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Birds

Birds are warm-blooded vertebrates whose covering of feathers is the one major characteristic that distinguishes them from all other animals. Birds have a four-chambered heart (shared with all mammals), forelimbs modified into wings (shared with bats), a calcareous-shelled egg, and keen vision, the major sense relied upon by birds for information about their environment. Their sense of smell is not highly developed, and auditory range is limited. Most birds are diurnal in habit.

In the scheme of biological classification, birds constitute the class Aves. There are approximately 8,700 living species, and more than 1,000 extinct species have been identified from fossil remains. For coverage of related topics in the *Macropædia* and the *Micropædia,* see the *Propædia,* section 313, and the *Index.*

This article is divided into the following sections:

BIRDS: THE CLASS AVES

GENERAL FEATURES

The smallest living bird is generally acknowledged to be the bee hummingbird of Cuba, which is 6.3 centimetres (2.5 inches) long and weighs less than 3 grams (about 0.1 ounce). The largest living bird is the ostrich, which may stand 2.5 metres (8 feet) tall and weigh 135 kilograms (300 pounds). Some extinct birds were even larger: the largest of the moas of New Zealand and the elephant birds of Madagascar may have reached over 3 metres (10 feet) in height. Among flying birds, the wandering albatross has the greatest wingspan, up to 3.5 metres (11.5 feet), and the trumpeter swan perhaps the greatest weight, 17 kilograms (38 pounds). A Pleistocene condorlike bird, *Teratornis incredibilis,* had an estimated wingspan of about 5 metres (16.5 feet) and was by far the largest known flying bird.

The ability to fly has permitted an almost unlimited radiation of birds, so that they are now found virtually everywhere on earth, from occasional stragglers over the polar ice caps to complex communities in tropical forests. In general the number of species found breeding in a given area is directly proportional to the size of the area and the diversity of habitats available. The total number of species is also related to such factors as the position of the area with respect to migration routes and wintering grounds of species that nest outside the area. In the United States, Texas and California have both the largest number of species recorded (545 and 461, respectively, including both resident and migrant species) and the largest number breeding (300 and 286). Seven hundred and seventy-five species, 650 of them breeding, have been recorded from North America north of Mexico. The figures for Europe exclusive of the Soviet Union are 577 and 420, and the figures for the Soviet Union are 704 and 622. Costa Rica, with an area of only about 51,000 square kilometres (about 20,000 square miles) and a known avifauna of at least 758 species, probably has the most diversified group for its size of any country.

BIRD AND HUMAN RELATIONSHIPS

Domestication

Wild birds and their eggs have been at least incidental sources of food for humans since their origin and still are in most societies. The eggs of some colonial seabirds, such as gulls, terns, and murres, or guillemots, and the young of some shearwaters (muttonbirds) are even now harvested in large quantities. With the development of agrarian human cultures, several species of birds became domesticated. Of these, chickens, ducks, geese, and pigeons, descended from the red jungle fowl (*Gallus gallus*), mallard duck (*Anas platyrhynchos*), greylag goose (*Anser anser*), and rock dove (*Columba livia*), respectively, were taken in early and have been selectively bred into many varieties. After the discovery of the New World, the turkey (*Meleagris gallopavo*), which had already been domesticated by the Indians, and the Muscovy duck (*Cairina moschata*) were brought to Europe and produced several varieties. Guinea fowl (*Numida meleagris*) from Africa were also widely exported and kept not only for food but because they are noisy when alarmed, thus warning of the approach of intruders. Besides being a food source, pigeons have long been bred and trained for carrying messages, and the ability of frigate birds to "home" to their nesting colonies has enabled inhabitants of the South Seas to send messages by these birds. For further information on domestication of birds, see FARMING: *Livestock farming* and *Poultry farming.*

With the development of modern culture, hunting evolved from a foraging activity to a sport, in which the food value of the game became secondary. Large sums are now spent annually on hunting waterfowl, quail, grouse, pheasants, doves, and other game birds. Sets of rules and conventions have been set up for hunting, and in one elaborate form of hunting, falconry, there is not only a large body of specialized information on keeping and training falcons but also a complex terminology, much of it centuries old.

Feathers have been used for decoration since early times. Their use in the headdresses of American Indians and various peoples of New Guinea is well known. Feather robes were made by Polynesians and Eskimos; down quilts, mattresses, and pillows are part of traditional European folk culture. Large feathers have often been used in fans, thereby providing an example of an object put to opposite uses—for cooling as well as for conserving heat. Whereas most feathers used in decorating are now saved as by-products of poultry raising or hunting, until early in the 20th century, egrets, grebes, and other birds were widely shot for their plumes alone. Ostrich farms have been established to produce plumes. Large quills were once widely used for writing, and feathers have long been used on arrows and fishing lures.

Many birds are kept as pets. Small finches and parrots are especially popular and easy to keep. Of these, the canary (*Serinus canaria*) and the budgerigar of Australia (*Melopsittacus undulatus,* sometimes called parakeet) are widely kept and have been bred for a variety of colour types. On large parks and estates, ornamental species like peafowl (*Pavo*) and various exotic waterfowl and pheasants are often kept. Zoological parks in many cities import birds from many lands and are a source of recreation for millions of people each year.

Birds as pests

With the rise of agriculture, man's relationship with birds became more complex. In regions where grain and fruit are grown, depredations by birds may be a serious problem. In North America various species of blackbirds (family Icteridae) are serious pests in grainfields; while in Africa a grain-eating finch, the red-billed quelea (*Quelea quelea*), occurs, like locusts, in plague proportions so numerous that alighting flocks may break the branches of trees. The use of city buildings for roosts by large flocks of starlings and blackbirds is also a problem, as is the nesting of albatrosses on airplane runways on Pacific islands. As a result of these problems, conferences on the control of avian pests are held with increasing frequency.

Although birds are subject to a great range of diseases and parasites, few of these are known to be capable of infecting man. Notable exceptions are ornithosis (or psittacosis), caused by one or more viruses that are transmitted directly to man from pigeons, parrots, and a variety of other birds, a serious and sometimes fatal disease resembling virus pneumonia. Encephalitis, an inflammation of the brain, is also serious and is transmitted from birds to man and to his domestic animals by biting arthropods, including mosquitos. Wild birds may also act as reservoirs for diseases that adversely affect domesticated birds. Much work has been done recently on the ecology of viruses, with more and more of them being found in birds.

The study of birds has contributed much to both the theoretical and practical aspects of biology. Darwin's studies of the Galápagos finches and other birds during the voyage of the "Beagle" were important in his formulation of the idea of the origin of species through natural selection. Study collections of birds in research museums still provide the bases for important studies of geographic variation, speciation, and zoogeography, because birds are one of the best known of animal groups. Early work on the domestic fowl added to the development of both genetics and embryology. The study of animal behaviour (ethology) has been based to a large extent on studies of birds by Konrad Lorenz, Nikolaas Tinbergen, and their successors. Birds also have been the primary group in the study of migration and orientation and the effect of hormones on behaviour and physiology.

Birds feature prominently in mythology and the literature of many countries. Some of their attributes, real or imagined, have led to their symbolic use in art as in language. The aesthetic and recreational pleasures of birdwatching are increasingly being recognized.

Man's impact on bird populations has become increasingly strong. Since 1680, approximately 80 species of birds have become extinct, and an even larger number are seriously endangered. While pollution and pesticides are important factors in the decline of certain large species, such as the peregrine falcon, osprey, and brown pelican, the destruction of natural areas and introduction of exotic animals and diseases have probably been the most devastating. Concerted efforts are required to ensure the survival of rare species and to learn as much as possible about them. (R.W.St.)

NATURAL HISTORY

Locomotion. Because of their body structure and their feathery body covering, birds are the best fliers among animals, better than the insects and the flying mammals, the bats. There are, however, considerable differences in flying ability among various birds. Penguins cannot fly but spend much of their time in the water swimming with their paddlelike wings; such birds as ostriches and kiwis have rudimentary wings and are permanently afoot. At the other extreme are the long-winged swifts and frigate birds that move from their perches only to fly, never to walk. Most birds alternate some walking or swimming with their flying.

Birds usually fly when they have any considerable distance to travel; there are exceptions, however. The mountain quail of California make their annual migrations up and down the mountains by foot. The murres, or guillemots, of the Greenland coast migrate southward by swimming; they begin their journey before the young have grown their flight feathers and before some of the adults at least have regrown their recently molted ones. The Adélie penguins may ride northward on drifting ice floes; at the approach of nesting time they swim back to the Antarctic continent and then walk over the ice to their breeding grounds many miles inland.

Flight. Birds fly by flapping their wings, steering mainly with their tails. A goshawk, pursuing its prey through the forest, uses its long tail in making quick turns, and the barn swallow uses its deeply forked tail in making the involved patterns of its graceful flight. Ducks with their short tails have a swift but direct flight. There is, however, such great diversity in birds' tails that the precise size and shape probably is not of critical importance.

Comparing a bird to an airplane, a bird's wing is both wing and propeller. The basal part of the wing supplies most of the supporting surface, the wing tip most of the propelling force. A bird's wing has many adjustable features: it can be shortened or lengthened by flexion; the feathers of the tip can be spread or closed; the angle of the whole wing or its parts, on one side or the other or on both sides, can be altered. All these adjustments make the aerodynamics of a bird's wing much more complicated than those of the airplane; consequently, the flight of a bird is much more varied and adaptable.

Flying ability varies widely among birds, and different types of wings correlate with different types of flight. Many songbirds use their short, rounded wings mostly to move with quick wing beats from perch to perch or from ground to perch. Ducks have pointed wings that, beaten at high speed, provide rapid flight for long distances. Swallows, terns, and frigate birds have long, pointed wings that enable these birds to fly and manoeuvre gracefully for hours with leisurely wing beats. Large herons with long, broad wings travel far with slow, measured wing beats, while buzzards soar high in the sky on their long, broad wings. Gulls and albatrosses with long, narrow wings sail along the beaches or over the waves with infrequent wing strokes. A hummingbird can whir its tiny wings so rapidly that it can hover as it thrusts its long bill into a blossom; it can even fly backward as it leaves the bloom.

Speed

The speed with which birds fly varies greatly from species to species, and of course individual birds can vary their speed. The data on the speed of birds' flight are difficult to evaluate. One of the complicating factors is that a bird's speed in relation to the ground may depend on the force of the wind. A bird flying at an airspeed of 40 mph with a 60-mph wind behind it would travel at 100 mph in relation to the ground (1 mile = 1.61 kilometeres). The same bird flying into a 60-mph wind would be losing ground at the rate of 20 mph. Despite the variables involved in determining a bird's speed of flight, the following generalized speeds, based on level flight in calm air, appear to be sound:

10–20 mph—many small songbirds such as sparrows and wrens

20–30 mph—many medium-sized birds such as thrushes and grackles, and larger, long-winged birds such as herons, pelicans, and gulls

20–40 mph—many small- and medium-sized birds such as starlings, chimney swifts, and mourning doves

40–60 mph—the faster-flying birds such as falcons, ducks, geese, and domestic pigeons

There are many faster records, often disputed, such as that of 200 mph for an Indian spine-tailed swift in level flight and 170 mph for a golden eagle in a dive. A homing pigeon has been timed at 94.3 mph.

Range

The record long-range flight of a bird species in a single season is undoubtedly held by the Arctic terns that migrate from a summering ground in the Arctic to a wintering ground in the Antarctic, travelling more than 11,600 kilometres (7,200 miles) one way. Some long-range flights are made very quickly: a blue-winged teal banded in Canada was recovered 6,100 kilometres (3,800 miles) away in Venezuela only 30 days later; a Manx shearwater, trapped at its nest in Wales and transported 5,200 kilometres (3,200 miles) to Massachusetts and released, returned home in 12½ days. Some very small birds regularly make long water crossings in a single flight. Ruby-throated hummingbirds fly across the more than 500-mile-wide Gulf of Mexico, and many warblers fly from the American coast to Bermuda, a journey of about the same distance. For further information on bird migration, see BEHAVIOUR, ANIMAL: *Animal migration.*

Flightlessness. Flight, so characteristic of birds, is maintained during the molt in most species by a gradual replacement of the flight feathers. However, ducks and geese, some rails and loons (divers), and auks shed all of their flight feathers at one time, immediately after the nesting season. Not until these feathers are replaced are the birds able to fly again. Most of these are birds that find their food by walking or swimming, as would be expected. Some ducks living in the marshes become very shy and retiring at this season, skulking in the reeds, but geese nesting in the Arctic barrens continue to walk about over the tundra, feeding. In another group of birds, however, the hornbills of Africa and Asia, only the females lose both flight and tail feathers at once; they stay in the nest until the feathers grow out again, being fed during this period by the males.

Evolutionary flightlessness

Some birds have completely lost the power of flight during the course of evolution. The close similarity in basic structure of both flightless and flying birds indicates, however, that they all had a common flying ancestor. The rudimentary wings and the flightless condition of the ostrich-like birds and the penguins is a secondary, specialized condition. That flightlessness is a secondary condition is made still more apparent in other flightless birds that belong to families most of whose members are capable of flight. The extinct great auk of the North Atlantic is one of the best known examples of such a flightless bird; the rail family also is noted for having many flightless species living on islands in the Pacific and the South Atlantic. Loss of flight seems to occur most often on isolated islands where there are no mammal predators. In New Zealand, where there are no native land mammals, not only are there many species of extinct flightless moas but also flightless kiwis, penguins, and rails and a duck, an owl, and several songbirds that are nearly flightless. The ostrich-like birds of continental distribution present an apparent contradiction to this correlation of mammal-free island habitats with bird flightlessness. Another adaptation, however, their great size, has enabled these forms to escape the predation of mammals.

Walking, hopping, and swimming. Terrestrial birds such as pheasants tend to walk; arboreal songbirds tend to hop as they travel from branch to branch. Parrots often walk along branches, and house sparrows hop when they come to the ground, while palm warblers walk on the ground and some songbirds, such as American robins and European blackbirds, may both walk and hop. Some birds with small feet, such as swifts, hummingbirds, bee eaters, and many hornbills, use their feet only for perching and rarely walk at all. Other birds with robust feet, such as guinea fowl and rails, do most of their moving about on foot. Jacanas with their greatly elongated toes and nails walk over floating water weeds, and herons with long legs wade in shallow water. The ostrich is probably the fastest running bird; some investigators have credited it with a speed of 50 mph (80.5 kph), at which time the length of its stride was about 25 feet (7.6 metres).

The usual position of a bird's body in walking is more or less parallel to the ground. But the penguins, with their feet far to the rear of their bodies, stand upright as they waddle along. When the Adélie penguin, however, makes its trek of many miles over the snow-covered ice to its breeding grounds, it may vary its awkward waddle with periods of tobogganing; *i.e.,* sliding along on its breast and propelling itself with thrusts of its feet.

Some water birds have become so adapted to swimming that they are practically helpless on land. In this class are loons, which shuffle awkwardly the few feet from the water to their nests. Swimming in birds is usually correlated with webbed feet, but coots and grebes, with only lobes on their toes, also swim and dive, and gallinules, without either webs or lobes on their toes, commonly swim. On the other hand, frigate birds, with partly webbed feet, never swim. Penguins swim through the water with their wings and use their webbed feet only for steering. Auks use their wings and webbed feet in swimming underwater.

Some birds such as the mallard usually swim at the surface, feeding only as far underwater as they can reach by dipping their heads. Other ducks, such as scoters and pochards, commonly dive to the bottom for their food, and cormorants, auks, and loons pursue fish underwater. Sometimes loons are taken at remarkable depths in fishermen's nets and on set lines, indicating that they may dive as deep as 200 feet.

Pond ducks, such as mallards and teals, spring straight up from the water's surface into the air in flight, but many swimming birds—for example, coots, grebes, cormorants, and diving ducks—take off with a long spattering run along the surface. (Au.L.R.)

Behaviour. Birds depend to a great extent on innate behaviour, responding automatically to specific visual or auditory stimuli. Even much of their feeding and reproductive behaviour is stereotyped. Feather care is vital to keep the wings and tail in condition for flying and the rest of the feathers in place where they can act as insulation. Consequently preening, oiling, shaking, and stretching movements are well developed and regularly used. Some movements, like the simultaneous stretching of one wing, one leg, and half the tail (all on the same side) are widespread if not universal among birds. Stretching both wings upward, either folded or spread, is another common movement, as is a shaking of the whole body beginning at the posterior end. Other movements have evolved in connection with bathing, either in water or in dust. Such comfort movements have frequently become ritualized as components of displays.

Many birds maintain a minimum distance between themselves and their neighbours, as can be seen in the spacing of a flock of swallows perched on a wire. In the breeding season most species maintain territories, defended areas ranging from the immediate vicinity of the nest to extensive areas in which a pair not only nests but also forages. The frequency of actual fighting is in birds greatly reduced by ritualized threat and appeasement displays. Birds range from solitary (*e.g.,* many birds of prey) to highly gregarious, like the guanay cormorants of the Peru Current off the west coast of South America, which nest in enormous colonies of hundreds of thousands and feed in large flocks with boobies and pelicans.

Vocalization

Auditory signals, like visual ones, are almost universal among birds. The most familiar vocalization of birds is that usually referred to as "song." It is a conspicuous sound (not necessarily musical) that is used, especially early in the breeding season, to attract a mate, to warn off another bird of the same sex, or both. As such it is usually associated with establishing and maintaining territories. Individual variation in songs of many species is well known, and it is believed that some birds can recognize their mates and neighbours by this variation. Many other types of vocalizations are also known. Pairs or flocks may be kept together by series of soft location notes. Alarm notes alert other individuals to the presence of danger; in fact, the American robin (and probably many other species) uses one note when it sees a hawk overhead and another when it sees a predator on the ground. Begging calls are important in stimulating parents to feed their young. Other calls are associated with aggressive situations, courtship, and mating. Nonvocal sounds are not uncommon. Some snipe and hummingbirds have narrow tail feathers that produce loud sounds when the birds are in flight, as do the narrowed outer primaries of the American woodcock. The elaborate courtship displays of grouse include vocalizations as well as stamping of the feet and noises made with the wings. Bill clapping is a common part of courtship in storks, and bill snapping is a common threat of owls.

Drawing by R. Keane based on M.H. Clench

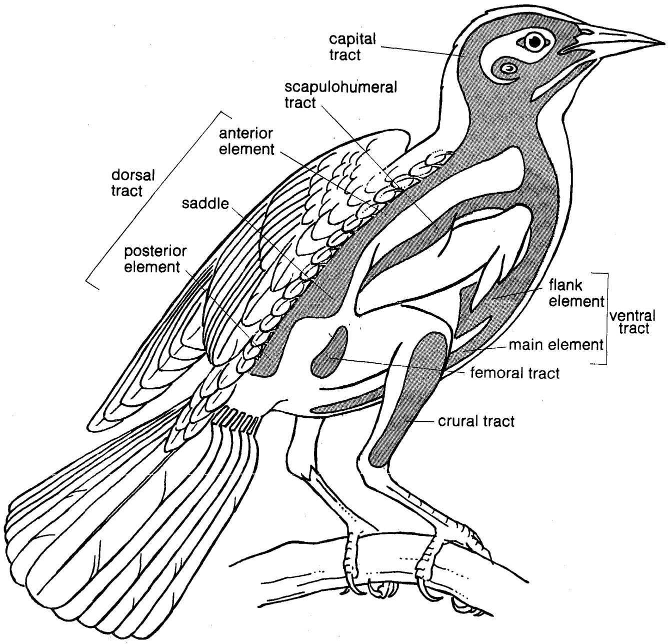

Figure 1: Basic body feather tracts on a generalized songbird. Shaded areas show the right half of each tract.

The nest

Most birds build nests in which the eggs are laid. Nests vary widely: they may be a scrape in the sand, a deep burrow, a hole in a tree or rock, an open cup, a globular or retort-shaped mass with a side entrance tube, or an elaborately woven hanging structure. The materials with which nests are made also vary widely. Some nests are lined with small stones, others are built of dirt or mud with or without plant material. Sticks, leaves, algae, rootlets, and other plant fibres are used alone or in combination. Some birds seek out animal materials such as feathers, horsehair, or snakeskin. The nest materials may be held together by weaving, sewing, or felting the materials themselves or with mud or spider webs. Swifts use saliva to glue nest materials together and to attach the nest to the supporting structure. In at least one species of swift, the entire nest is made of saliva and is the prized ingredient of birds' nest soup in the Orient. All birds incubate their eggs, except megapodes (mound builders), which depend on the heat generated by decaying vegetation or other external sources, and brood parasites, which lay their eggs in the nests of other species. Murres and the king and emperor penguins build no nest but incubate with the egg resting on top of the feet. In most birds a brood patch is developed. This bare area on the abdomen is edematous (fluid filled) and highly vascularized and is in direct contact with the eggs during incubation. Its development during the breeding season is under hormonal control. When the parent is off the nest, adjacent feathers are directed over the brood patch, and it is usually not apparent. A few birds (*e.g.,* boobies) keep their webbed feet over the eggs during incubation.

Incubation takes from 11 to 80 days, depending at least in part on the size of the bird and the degree of development at hatching. Most songbirds and members of some other groups are hatched nearly naked and helpless (altricial) and are brooded until well able to regulate their body temperature. They are fed by the parents until after they are capable of flight. The young of numerous other birds, such as chickens, ducks, and shorebirds, are hatched with a heavy coat of down and are capable of foraging for themselves almost immediately (precocial). Still others, such as the petrels and the auks are downy when hatched but remain in the nest and are fed by their parents.

The length of time parents care for young birds varies widely. Young megapodes can fly shortly after hatching and are entirely independent of their parents; young royal albatrosses may spend up to 243 days at the nest and in the area immediately around it before they can fly. The length of time needed to attain independence is related to size and condition at hatching. Ground-nesting birds tend to take less and hole-nesting birds more time than the average.

The number of eggs in a set varies from 1 to about 20. Some species invariably lay the same number per clutch (determinate laying), whereas in the majority the number is variable (indeterminate laying). In species of the latter category, clutch size tends to be smaller in tropical regions than in cold ones. There is also a tendency for birds in warm regions to make more nesting attempts in a given season. In the Arctic, where the season is very short, the cycle of breeding and the molt that follows it are telescoped into a minimum of time.

Feeding habits. The earliest birds were probably insectivorous, as are many modern ones, and the latter have evolved many specializations for catching insects: swifts, swallows, and nightjars have wide gapes for catching insects on the wing; some woodpeckers can reach wood-boring grubs while others can catch ants by probing anthills with their long, sticky tongues; thrashers dig in the ground with their bills; tree creepers and woodhewers probe bark crevices; and warblers glean insects from many kinds of vegetation. Raptorial birds have evolved talons and hooked bills for feeding on larger animals, and vultures have bare heads and tearing bills for feeding on carrion. Herons have spearlike bills and trigger mechanisms in the neck for catching fish, while kingfishers, terns, and boobies plunge into the water after similar prey. Long-billed waders probe for worms and other invertebrates. Of the many kinds of birds that feed on plant material, most use seeds, fruit, or nectar, which are high in food value; leaves and buds are eaten by fewer species. While some kinds of birds feed entirely on a single kind of food, others may take a wide range of foods, and many have seasonal changes.

FORM AND FUNCTION

Body proportions. Birds arose as warm-blooded, arboreal, flying animals with forelimbs adapted for flight and hindlimbs for perching. This basic plan has become so modified, through the course of evolution, that in some forms it is difficult to recognize. The maximum size attainable by flying birds is limited by the fact that wing area varies as the square of linear proportions, and weight or volume as the cube. On the other hand, the minimum size is probably governed by another aspect of the surface-volume ratio: the relative increase, with decreasing size, in surface through which heat can be lost. The largest flying birds have highly pneumatic skeletons (part of the bone is replaced by air cavities) and other adaptations for reducing weight; the small size of some hummingbirds may be facilitated by the decrease in heat loss resulting from their becoming torpid at night.

When birds lose the power of flight, the limit on their maximum size is lifted, as can be seen in the ostrich and other ratite birds. Some birds (auks, diving petrels, and certain ducks) use the wings for propulsion underwater as well as in the air. When birds that "fly" underwater lose the ability to fly in air, the wings become highly modified as paddles, as in the penguins.

Wing shape

The types of flight found in birds vary considerably. At least two major types of modifications for gliding or soaring are found. The albatrosses and some other seabirds have long, narrow wings and take advantage of winds over the oceans, whereas some vultures and hawks have broad wings with slotted tips and make more use of updrafts and winds deflected by hills. Short, broad wings are characteristic of chicken-like birds, which fly up with a rush of rapid wing beats. Birds like ducks, pigeons, and falcons, which fly rapidly with continuous wing beats, tend to have moderately long, pointed wings, while swifts and hummingbirds, with their narrow, curved wings fly rapidly and manoeuvre easily. The shape of a bird's tail also appears to be related to flight. Forms with deeply forked tails, such as frigate birds, terns, and some swallows, manoeuvre easily, whereas the opposite extreme, long, graduated tails, are often found in rapid, direct fliers such as some parrots and doves. Woodpeckers and some other climbing birds have strong tail feathers with stout shafts, which they use as props while on the trunks of trees.

The bipedal gait, dictated by modification of the forelimbs for flight, necessitates manipulating food by the bill and feet and poses problems in balance. The relative lengths of the segments of the legs must be such that as the bird shifts from a standing to a sitting position, its centre of gravity remains over the feet. As some birds moved out of the trees and became terrestrial or aquatic, their legs were accordingly modified. The toes became shorter and the opposable first toe was lost in rapidly running forms like rheas and ostriches, and the toes became very long in birds that walk on aquatic vegetation or soft ground. In very large, slow-moving birds such as moas, the leg bones became very heavy. Wading birds developed long legs, and climbing birds developed short legs with strongly curved, sharp claws. In swimming and diving birds, webs developed between the toes or lobes on the sides of the toes.

Feathers and molt. Feathers are unique to birds and characteristic of them. Like the scales of reptiles, and those on the feet of birds, feathers are made of keratin, a fibrous protein also found in hair. Feathers vary considerably in structure and function (Figure 2). Contour feathers form most of the surface of the bird, streamlining it for flight and often waterproofing it. The basal portion may be downy and thus act as insulation. The major contour feathers of the wing (remiges) and tail (rectrices) and their coverts function in flight. Contour feathers grow in tracts (pterylae) separated by bare areas (apteria) and develop from follicles in the skin.

Drawing by R. Keane

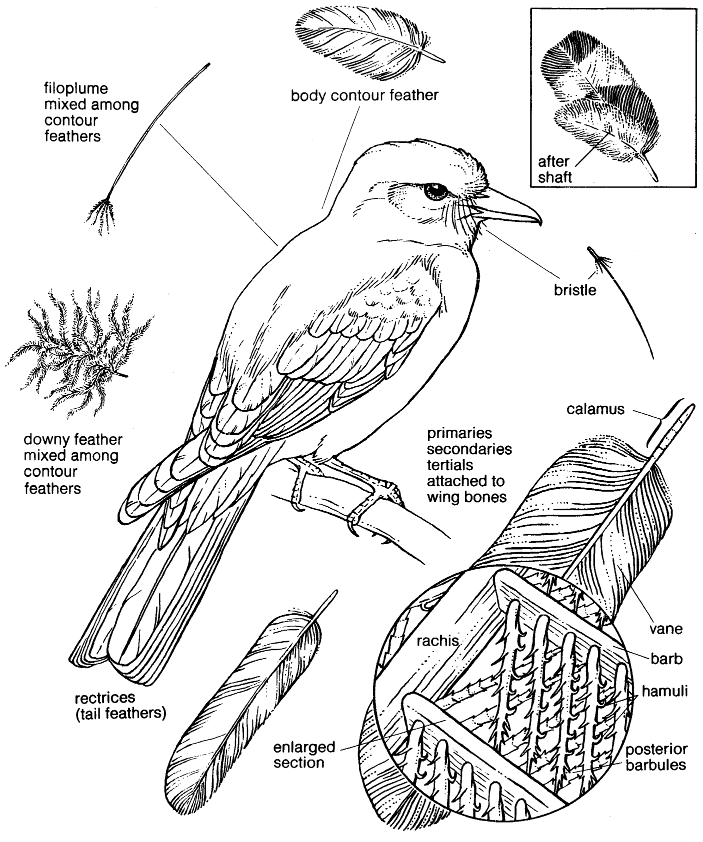

Figure 2: Feather types and their distribution on a typical perching bird.

Feather structure

The typical contour feather consists of a tapered central shaft, the rachis, with paired branches (barbs) on each side. An unbranched basal section of the rachis is called the calamus, part of which lies beneath the skin. The barbs, in turn, have branches, the barbules. The barbules on the distal side of each barb have hooks (hamuli) that engage the barbules of the next barb. The barbs at the base of the vane are often plumaceous; *i.e.,* lacking in hamuli and remaining free of each other. In many birds each contour feather on the body (but rarely on the wings) is provided with a complex branch, the aftershaft, or afterfeather, that arises at the base of the vane. The aftershaft has the appearance of a second, smaller feather, growing from the base of the first. Down feathers have loose-webbed barbs, all rising from the tip of a very short shaft. Their function is insulation, and they may be found in both pterylae and apteria in adult birds. They also constitute the first feather coat of most young birds. Filoplumes are hairlike feathers with a few soft barbs near the tip. They are associated with contour feathers and may be sensory or decorative in function. Bristle-like, vaneless feathers occur around the mouth, eyes, and nostrils of birds. They are especially conspicuous around the gape (corners of the mouth) of birds that catch insects in the air. Some bristles function as eyelashes on ground-dwelling birds, and the bristles over the nostrils may serve as filters.

The contour feathers are shed and replaced (molted) at least once a year, usually just after the breeding season. In addition, many birds have at least a partial molt before the breeding season. A typical series of molts and plumages would be juvenile plumage, postjuvenile (also called first prebasic) molt, first winter (or first basic) plumage, first prenuptial (or pre-alternate) molt, first nuptial (or alternate) plumage, first postnuptial (first annual, or second prebasic) molt, second winter (or basic) plumage, etc. Molt of the remiges and rectrices usually occurs as part of the annual molt and can be serial, from the innermost feather out (centrifugal), from the outermost in (centripetal), or simultaneous. Normally it is symmetrical between the right and left sides.

Coloration

Colour in birds is caused by pigments or structure. Buffs, red browns, dark browns, and blacks are caused by melanins, pigments synthesized by the bird and laid down in granules. Yellows, oranges, and reds come from carotenoid or lipochrome pigments; these originate at least in part from the food and are diffused in the skin and feathers. Porphyrin feather pigments occur in birds but less frequently than melanins and carotenoids. Blue colours in feathers are structural, based on a thin, porous layer of keratin overlying melanin pigment. Most greens result from the addition of yellow pigment to the structural blue colour. Iridescent colours result from thinly laminated structure of the barbules and are enhanced by underlying melanin deposits.

Birds' feet are covered with scales like those of reptiles. The scales are occasionally shed, but the timing of this molt is not known. The toes are tipped with claws, and vestigial claws are not infrequently found on the tips of the first two digits of the wing.

The bill is covered with a sheet of keratin, the rhamphotheca, which in petrels and a few other birds is divided into plates. In birds that probe for food (kiwis, woodcock, etc.), many sensory pores are found near the tip of the bill. Both melanins and carotenoids are found in the rhamphotheca and in the scales of the feet.

The skin of a bird is almost without glands. The important exception is the oil (uropygial) gland, which lies on the rump at the base of the tail. The secretion of

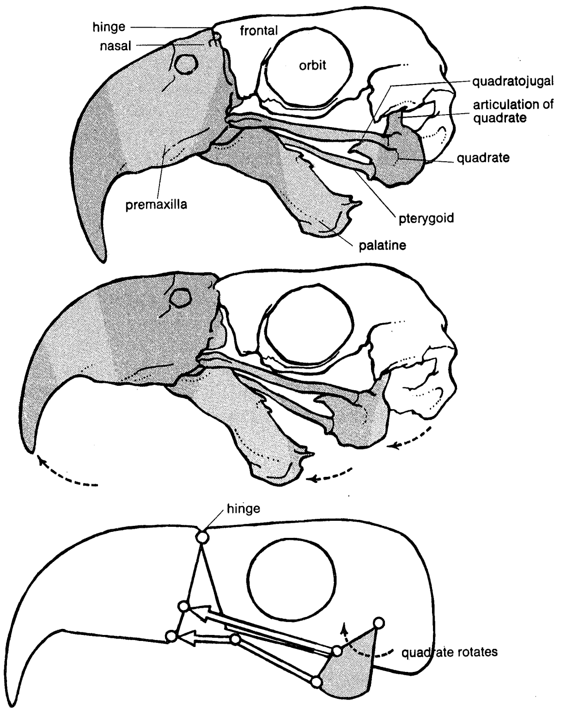

Figure 3: *Kinesis of the cranium of a macaw.*
(Top) With upper mandible lowered. (Centre) With upper mandible raised. (Bottom) Showing forces acting on mandible.

Drawing by R. Keane based on A. Bellairs and C.R. Jenkin, "The Skeleton of Birds," in J. Marshall (ed.), *Biology and Comparative Physiology of Birds,* vol. 1; Academic Press, Inc.

this gland contains approximately one-half lipids (fats) and is probably important in dressing and waterproofing the plumage. In a few birds, the secretion has a strong, offensive odour. Some birds, in which the oil gland is small or absent, have a specialized type of feather (powder down) that grows continuously and breaks down into a fine powder, believed to be used in dressing the plumage.

Movable upper mandible

Skeleton. The avian skeleton (Figure 4) is notable for its strength and lightness, achieved by fusion of elements and by pneumatization (*i.e.,* presence of air cavities). The skull represents an advance over that of reptiles in the relatively larger cranium with fusion of elements, made possible by the fact that birds have a fixed adult size. Birds differ from mammals in being able to move the upper mandible, relative to the cranium. When the mouth is opened, both lower and upper jaws move: the former by a simple, hingelike articulation with the quadrate bone at the base of the jaw, the latter through flexibility provided by a hinge between the frontal and nasal bones. As the lower jaw moves downward, the quadrate rocks forward on its articulation with the cranium, transferring this motion through the bones of the palate and the bony bar below the eye to the maxilla, the main bone of the upper jaw.

The number of vertebrae varies from 39 to 63, with remarkable variation (11 to 25) within the cervical (neck) series. The principal type of vertebral articulation is heterocoelous (saddle shaped). The three to 10 (usually five to eight) thoracic (chest) vertebrae each normally bear a pair of complete ribs consisting of a dorsal vertebral rib articulating with the vertebra and with the ventral sternal rib, which in turn articulates with the sternum (breastbone). Each vertebral rib bears a flat, backward-pointing spur, the uncinate process, characteristic of birds. The sternum, ribs, and their articulations form the structural basis for a bellows action, by which air is moved through the lungs. Posterior to the thoracic vertebrae is a series of 10 to 23 fused vertebrae, the synsacrum, to which the pelvic girdle is fused. Posterior to the synsacrum is a series of free caudal (tail) vertebrae and finally the pygostyle, which consists of several fused caudal vertebrae and supports the tail feathers. The sternum consists of a plate lying ventral to the thoracic cavity and a median keel extending ventrally from it. The plate and keel form the major area of attachment for the flight muscles. The bones of the pectoral girdle consist of the furcula (wishbone) and the paired coracoids and scapulas (shoulder blades). The sword-shaped scapula articulates with the coracoid and humerus (the bone of the upper "arm") and lies just dorsal to the rib basket. The coracoid articulates with the anterior (forward) edge of the sternum and with the scapula, humerus, and furcula. The furcula connects the shoulder joints with the anterior edge of the keel of the sternum. It consists of paired clavicles (collarbones) and, probably, the median, unpaired interclavicle.

Wing skeleton

The bones of the forelimb are modified for flight with feathers. Major modifications include restricting the motion of the elbow and wrist joints to one plane, reduction of the number of digits, loss of functional claws, fusion of certain bones of the "hand" (the metacarpals and most of the carpals) into a carpometacarpus, and modification of the elements, especially those toward the tip of the limb (distal), for the attachment of feathers. The wing bones are hollow, and the cavity in the humerus, at least, is connected with the air-sac system. As a general rule, large flying birds have proportionally greater pneumaticity in the skeleton than small ones. The highly pneumatic bones of large flying birds are reinforced with bony struts at points of stress. The humerus, radius, and ulna are well developed. The secondary flight feathers are attached to the ulna, which thus directly transmits force from the flight muscles to these feathers and is therefore relatively heavier than the radius. Two small wrist bones are present: the radiale, or scapholunar, and the ulnare, or cuneiform. The former lies between the distal end of the radius and the proximal part (the part toward the body) of the carpometacarpus. When the elbow joint is flexed (bent), the radius slides forward on the ulna and pushes the radiale against the carpometacarpus, which in turn flexes the wrist. Thus the two joints operate simultaneously. The U-shaped ulnare articulates with the ulna and the carpometacarpus. Anatomists differ on which bones of the reptilian "hand" are represented in the bird's wing. Embryological evidence suggests that the digits are II, III,

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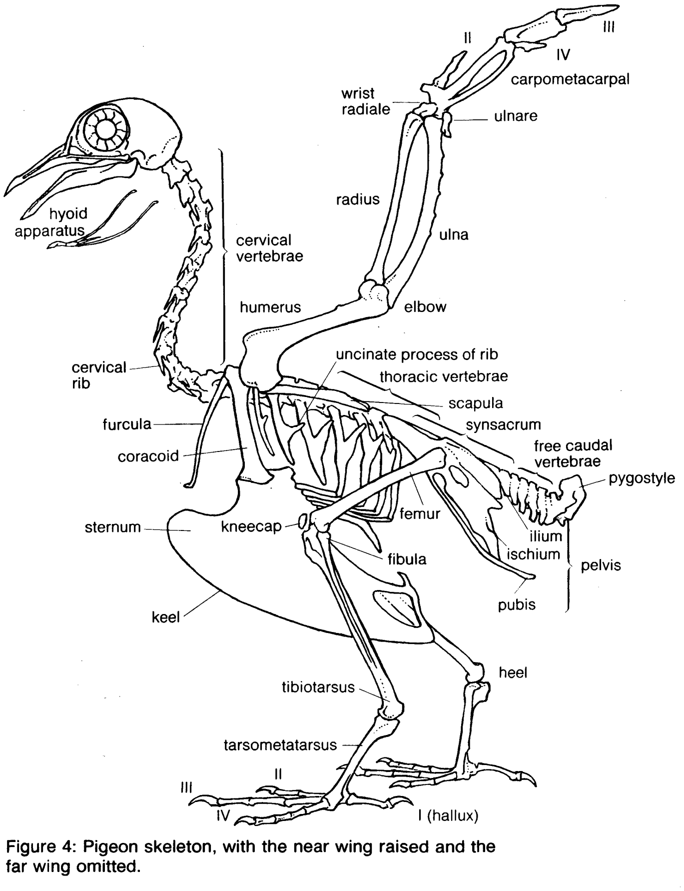

Figure 4: Pigeon skeleton, with the near wing raised and the far wing omitted.

and IV, but it is possible that they are actually I, II, and III. The carpometacarpus consists of fused carpals (bones of the wrist) and metacarpals (bones of the palm), metacarpals II and III (or III and IV) contributing the greater part of the bone. The phalanges (bones of the "fingers") are reduced to one each on the outer and inner digits and two on the middle one. The primary flight feathers are attached to the carpometacarpus and digits, the number attached to each being characteristic of the various major groups of birds.

Pelvis and leg

The pelvic girdle consists of three paired elements, the ilia, ischia, and pubes, which are fused into a single piece with the synsacrum. The ilium is the most dorsal element and the only one extending forward of the acetabulum (the socket of the leg). The ilium is fused with the synsacrum and the ischium, the latter of which is fused with the pubis. All three serve as attachments for leg muscles and contribute to the acetabulum, which forms the articulation for the femur. The leg skeleton consists of the femur (thighbone), tibiotarsus (main bone of the lower leg), fibula, tarsometatarsus (fused bones of the ankle and middle foot), and phalanges (toes). The fibula is largest at its proximal (upper) end, where it forms part of the knee joint and tapers to a point distally, never forming part of the ankle joint. The latter joint is simplified, there being but two bones involved: the tibiotarsus, consisting of the tibia (the so-called shinbone in man) fused with the three proximal tarsals (upper ankle bones), and the tarsometatarsus, resulting from the fusion of metatarsals I through IV and the distal row of tarsals. Metatarsals II through IV contribute most to the tarsometatarsus. The basic number of phalanges (sections) on the toes is two, three, four, and five, respectively; *i.e.,* one more than the number of the toe. Most birds have four toes, the fifth being always absent, but there are many variations in the number of digits, or phalanges, representing reductions of the basic arrangement.

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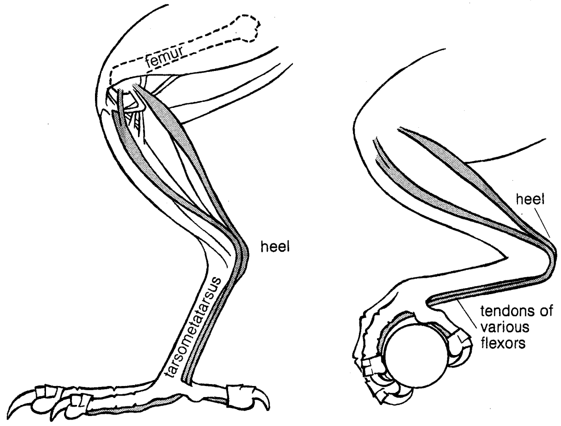


Figure 5: Perching mechanism of a pigeon with the leg extended and flexed.

The basic avian foot is adapted for perching. The first, or hind, toe (hallux) opposes the other three, and the tendons for the muscles that bend the toes pass behind the ankle joint in such a way that when the ankle is bent the toes are also. The weight of a crouched bird thus keeps the toes clasped around the perch.

Muscles

Internal organs. The cardiac (heart) muscles and smooth muscles of the viscera of birds resemble those of reptiles and mammals. The smooth muscles in the skin include a series of minute feather muscles, usually a pair running from a feather follicle to each of the four surrounding follicles. Some of these muscles act to raise the feathers, others to depress them. The striated (striped) muscles that move the limbs are concentrated on the girdles and the proximal parts of the limbs. Two pairs of large muscles move the wings in flight: the pectoralis, which lowers the wing, and the supracoracoideus, which raises it. The latter lies in the angle between the keel and the plate of the sternum and along the coracoid. It achieves a pulley-like action by means of a tendon that passes through the canal at the junction of the coracoid, furcula, and scapula and attaches to the dorsal side of the head of the humerus. The pectoralis lies over the supracoracoideus and attaches directly to the head of the humerus. In most birds the supracoracoideus is much smaller than the pectoralis, weighing as little as one-twentieth as much; in the few groups that use a powered upstroke of the wings (penguins, auks, swifts, hummingbirds, and a few others), the supracoracoideus is relatively large. Avian striated muscles contain a respiratory pigment, myoglobin. There are relatively few myoglobin-containing cells in "white meat," whereas "dark meat" derives its characteristic colour from their presence. The former type of muscle is used in short, rapid bursts of activity, whereas the latter is characteristic of muscles used continuously for long periods and especially in muscles used during diving.

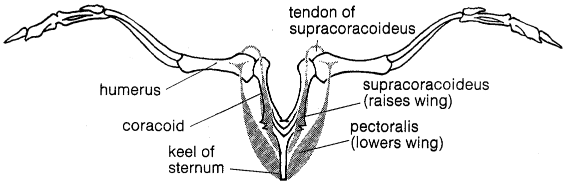


Figure 6: Pectoral girdle of a generalized bird.

Drawing by R. Keane based on L. Darling and L. Darling, *Bird* (1962); Houghton Mifflin Company

Circulatory features

The circulatory system of birds is advanced over that of reptiles in several ways: (1) there is a complete separation between the pulmonary (lung) and systemic (body) circulations, as in the mammals; (2) the left systemic arch (aortic artery) is lost, blood passing from the heart to the dorsal aorta via the right arch; (3) the postcaval vein is directly connected with the renal portal that connects the kidneys with the liver; and (4) the portal circulation through the kidneys is greatly reduced. Birds' hearts are large—0.2 to over 2.4 percent of body weight, as opposed to 0.24 to 0.79 percent in most mammals.

The avian lung differs from the type found in other land vertebrates, in containing fine tubes (capillaries) through which air passes and through the walls of which gas ex-

Drawing by R. Keane based on A.L. Thomson (ed.), *A New Dictionary of Birds;* British Ornithologists' Union and J. Marshall (ed.), *Biology and Comparative Physiology of Birds* (1961); Academic Press

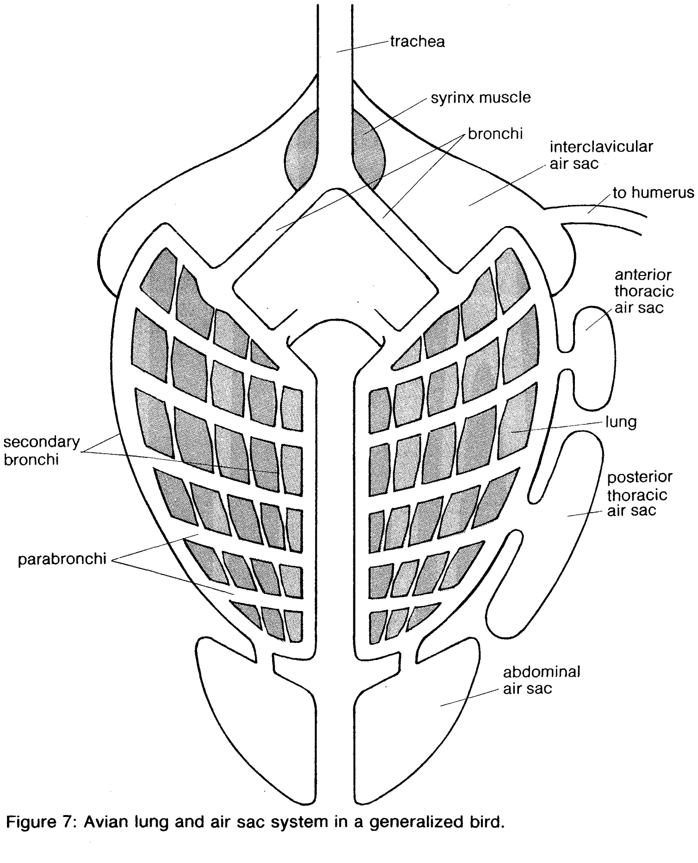


Figure 7: Avian lung and air sac system in a generalized bird.

change takes place. Several pairs of nonvascular air sacs are connected with the lungs and extend into the pneumatic parts of the skeleton. The sound-producing organ in birds is the syrinx, located where the trachea (windpipe) divides into the bronchial tubes. The sounds are made by the flow of air setting up vibrations in membranes formed from part of the trachea, bronchi, or both. Muscles between the sternum and trachea or along the trachea and bronchi vary tension on the membranes.

The avian digestive system shows adaptations for a high metabolic rate and flight. Enlargements (collectively called the crop) of parts of the esophagus permit the temporary storage of food prior to digestion. The stomach is typically divided into a glandular proventriculus and a muscular gizzard, the latter lying near the centre of gravity of the bird and compensating for the lack of teeth and the generally weak jaw musculature. Otherwise, the digestive system does not vary markedly from the general vertebrate type.

Like reptiles, birds possess a cloaca, a chamber that receives digestive and metabolic wastes and reproductive products. A dorsal outpocketing of the cloaca, the bursa of Fabricius, controls antibody-mediated immunity in young birds. The bursa regresses with age, and thus its presence or absence may be used to determine age.

The testes of the male bird are internal, like those of reptiles. Intromittent organs are found in only a few groups (waterfowl, cracids, tinamous, ratites). The distal part of the vas deferens (the seminal sac) becomes enlarged and convoluted in the breeding season and takes on both secretory and storage functions. In passerine birds, at least, this enlargement and the adjacent part of the cloaca form a cloacal protuberance, a swelling visible on the outside of the bird. Usually only the left ovary and oviduct are functional. The albumen, membranes, and shell are laid down in the oviduct as the egg moves down it. The gonads and accessory sexual organs of both sexes enlarge and regress seasonally. In the breeding season, the testes of finches may increase over 300-fold in volume over their winter size.

Temperature control

Birds are homeothermic (warm-blooded) and maintain a body temperature of approximately 41° C (106° F). This temperature may be 1–1.7° C less during periods of sleep and up to 2° C higher at times of great activity. Feathers, including down, provide effective insulation. In addition, layers of subcutaneous fat add further insulation in penguins and in some other water birds. Reduction of heat loss from the feet in cold weather is accomplished by reducing blood flow to the feet and by a heat-exchange network in the blood vessels of the upper leg, so that the temperature of blood flowing into the unfeathered part of the leg is very low.

Birds differ from mammals in lacking sweat glands, hence heat loss is accomplished by rapid panting, which reaches 300 respirations per minute in domestic hens. Some heat dissipation can be accomplished by regulation of blood flow to the feet. In hot climates, overheating is often prevented or reduced by behavioural means—by concentrating activities in the cooler parts of the day and seeking shade during the hot periods. Temporary hypothermia (lowered body temperature) and torpor are known for several species of nightjars, swifts, and hummingbirds. Torpor at night is believed to be widespread among hummingbirds. The heart rate of birds varies widely—from 60 to 70 beats per minute in the ostrich to more than 1,000 in some hummingbirds.

The kidneys lie in depressions that are located on the underside of the pelvis. The malpighian bodies, which are the active tubules of the kidney, are very small in comparison to those of mammals, ranging from 90 to 400 per cubic millimetre. More than 60 percent of the nitrogen is excreted as uric acid or its salts. There is some resorption of water from the urine in the cloaca, with uric acid remaining. There is no urinary bladder, the urine being voided with the feces. In marine birds, salt is excreted in a solution from glands lying above the eyes through ducts leading to the nasal cavity.

EVOLUTION AND PALEONTOLOGY

Archaeopteryx

The earliest known fossil bird is *Archaeopteryx lithographica*, which was discovered in Upper Jurassic deposits in Bavaria. This bird was about the size of a magpie. It resembled some reptiles, however, and differed from Recent birds in many ways: (1) the jaws contained teeth set in sockets; (2) the articulations between the vertebrae were amphicoelous (concave at both ends); (3) there were only six sacral vertebrae; (4) the long tail was made up of a series of free vertebrae each bearing a pair of rectrices; (5) the slender ribs lacked articulations with the sternum and uncinate processes (flat upward projections); (6) ventral ribs (gastralia) were present behind (posterior to) the sternum; (7) the sternum was short and not keeled; (8) the bones were not pneumatic; (9) the third metacarpal bone in the wing was fused to the carpals, but the first two metacarpals were free, resulting in three movable digits of the "hand," all with functional claws; (10) the fibula was as long as the tibia; (11) the metatarsal bones were free; (12) the cerebral hemispheres were elongated and slender, and the cerebellum lay behind the midbrain, not overlapping it from behind or crowding it downward. The avian characteristics of *Archaeopteryx* included the possession of feathers, the elongated, backward-directed pubis, the furcula, and the opposable hallux. In the structure of the beak, eye, and jaw articulation, in the fusion of the third metacarpal with the carpals, and in the fusion of each of the distal tarsals with the corresponding metatarsal, *Archaeopteryx* represented an intermediate stage between reptiles and modern birds.

The absence of a keel on the short sternum indicates that *Archaeopteryx* did not fly but glided. The opposable hallux, indicative of the perching type of foot, and the clawed digits of the hand point to an arboreal existence. From the arrangement of feathers on the wing and the number and arrangement of bones in the limbs, it appears that *Archaeopteryx* was near the main line of avian evolution. From the fact that the skull was diapsid (*i.e.,* had two "windows") and from certain features of the limb bones,

Drawing by R. Keane

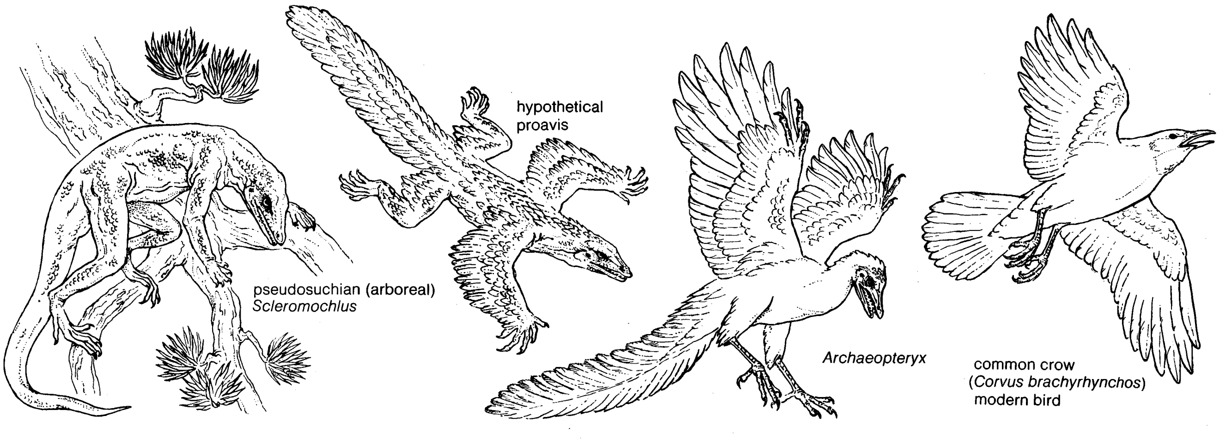


Figure 8: Four stages in the evolution of modern birds.

it appears that *Archaeopteryx* was descended from reptiles of the Triassic order Thecodontia.

The origins of flight

By the Triassic Period (225,000,000 years ago) a group of small bipedal reptiles, the pseudosuchians, were well established. Their skulls had much in common with that of *Archaeopteryx,* although they had heavier jaws and smaller eyes. It is likely that one group of pseudosuchians became arboreal. The advantages of such a life would be safety from large terrestrial predators and an abundance of insect food. Once these reptiles were in the trees, selective pressures would favour mutations leading to many avian features. The swaying of branches would favour the evolution of the grasping foot. Use of the forelimbs in climbing from branch to branch would favour enlargement of the claws and elongation of the forelimb, which was short in the bipedal ancestors. Greater visual acuity and more effective coordination are of special advantage in arboreal animals. Natural selection thus favoured larger eyes, narrower snouts (permitting better forward vision), and greater development of the cerebral lobes and cerebellum of the brain. The smaller jaws may also indicate the advantages of lightness, balance, and a specialization for feeding on insects as opposed to the apparently more general carnivorous diet of the terrestrial ancestors. Perhaps most important was the development of homeothermy (internal temperature control). A warm-blooded insect eater has an enormous advantage in being able to capture insects when they are cold and slow to react. It is also advantageous in the wind-moved environment of the treetops. In addition to increased food intake and advanced respiratory and circulatory systems, however, homeothermy requires effective insulation. It is likely that feathers evolved to fill this requirement, although many authorities believe the origin of feathers was directly connected with flight. Just how feathers evolved from reptilian scales is unknown, but it is known that the two are similar in chemical composition and that some pseudosuchians had scales bearing an imprint of a feather-like pattern on their surface. Elongated feathers on the forelimb and tail may have evolved for balancing and for gliding to produce the *Archaeopteryx* stage.

In the evolution of modern birds from an *Archaeopteryx*-like form, the development of active flight must have occurred early. This meant an increase in size of the muscles moving the wing and the development of a keel on the sternum as an added area of attachment for these muscles. As the tail took on more of a steering function and less of a supportive one, it became shorter and more readily moved as a unit. Feathers became increasingly specialized for different functions, and at the same time, the trends in the development of the eyes, brain, and respiratory and circulatory systems associated with the evolution of the homeothermic, arboreal, gliding types continued. By the time birds became strong fliers, they were ready to radiate out into many new environments; and by the Cretaceous Period (136,000,000 to 65,000,000 years ago), they had begun to do so. This radiation has produced the large array of adaptive types known today.

The lightness and pneumaticity of bird bones makes them poor candidates for fossilization. As might be expected, heavyboned diving birds and large flightless birds are disproportionally represented in the record.

Cretaceous birds

One of the best known groups of fossil birds consists of *Hesperornis* and its relatives. These birds were highly specialized foot-propelled divers of the Upper Cretaceous. The known species of *Hesperornis* were up to six feet (1.8 metres) long and had completely lost the power of flight. The sternum lacked a keel; the humerus was small and weak; and the other, more distal, elements of the wing were missing. The pelvis and hindlimb had a strong but superficial resemblance to those of modern loons and grebes—the pelvis was narrow; the femur short and stout with a hingelike articulation with the pelvis; the tibiotarsus long, with a long cnemial crest (a projection at its upper end); and the tarsometatarsus laterally compressed. Two major features (and several less obvious ones) indicate, however, that the resemblance was the result of convergent evolution: the ischia and pubes were free for most of their length, and the cnemial process was made up entirely of the patella; in the loon, this process is derived from the tibiotarsus. *Hesperornis* was remarkable for three features: it had teeth set in grooves, not sockets, in the maxilla and mandible; the phalanges of the stout fourth toe had a unique rotary ball-and-flange type of articulation; and the free tail vertebrae had broad lateral projections and limited vertical motion, indicating that the tail was somewhat beaver-like in its action. *Baptornis,* a contemporary relative of *Hesperornis,* was smaller and less strongly modified. While flightless, it had less reduced wings than *Hesperornis,* and it lacked the peculiar modifications of the fourth toe and caudal vertebrae.

Living on the same seas as *Hesperornis* and *Baptornis* was a group of flying birds known as *Ichthyornis* and *Apatornis.* Although not related to gulls, these birds resembled them superficially and may well have been their ecological counterparts. It was long believed that *Ichthyornis* had teeth, like *Hesperornis;* but it is now thought that the toothed jaws formerly thought to belong to *Ichthyornis* were really those of a small mosasaur, a marine reptile.

Large flightless predators

After the extinction of the dinosaurs and before large carnivorous mammals evolved, two groups of large flightless birds evolved to fill a similar niche. From the upper Paleocene to the middle Eocene, *Diatryma* and its relatives were major predators in the Northern Hemisphere. The largest species stood over two metres (seven feet) tall and had stout hooked beaks. They are of uncertain relationships but may have been distantly related to cranes and rails. The second group, that of *Phororhacos* and related genera, had a long history (from the lower Oligocene to the middle Pliocene) in South America, which was without large carnivores until relatively late. Fragmentary Pleistocene material from Florida has also been assigned to this group. The *Phororhacos* line evidently evolved from cariama-like stock and radiated into numerous genera and species, the largest of them (*Onactornis*) standing 2.5 metres (eight feet) tall and having a skull 80 centimetres (31 inches) long and 40 centimetres (16 inches) high.

Large grazing or browsing birds appear to have evolved several times. On continents where there are large predators, these birds have always been rapid runners (ostriches, rheas, emus), but on islands lacking such predators, they were slow-moving, heavy-bodied birds. Two such groups were the elephant birds of Madagascar and the moas of New Zealand, the largest in each group approaching 10 feet in height. Fragmentary fossil material from Eocene and Oligocene deposits in Egypt indicates that similarly adapted birds occurred there before the advent of large carnivores.

Except for the *Hesperornis* line, teeth appear to have been lost very early in the history of birds, but fish-eating birds have evolved several toothlike structures for grasping their prey. Perhaps the most remarkable adaptation was that of *Osteodontornis* and its relatives, large, flying marine birds that flourished from the lower Eocene to the Miocene. In these birds, there were bony projections of the upper and lower jaws, which were covered by the ramphotheca, forming sharp, toothlike structures.

The fact that fewer bird fossils are found in earlier deposits is well illustrated by expressing the number of known species in a given geological period in terms of the duration of the period. About 35 species of birds are known from Cretaceous deposits, which were laid down over an estimated 71,000,000 years, giving a figure of 0.5 species per 1,000,000 years. The corresponding figure for the Paleocene is 1.2 (12 species in 10,000,000 years) and that for the Eocene, 5.4 (87 species in 16,000,000 years). Up-to-date figures for the later periods are not available, but estimates based on several recent sources are 7.9 per 1,000,000 years in the Oligocene, 12.3 in the Miocene, and 21.8 in the Pliocene. The Pleistocene, which lasted approximately 2,500,000 years, has yielded nearly a thousand species of fossil birds. From this it is evident that very little is known about the early avifaunas. It is known, however, that, as might be expected, the birds in the earlier periods differ more from Recent species than do those of the later periods. Of the 12 families of birds recorded from Cretaceous deposits, only two are still extant, whereas the majority of species recorded from the Pleistocene were structurally little, if at all, different from living forms. Thus the absence of a group in the fossil record, especially in the earlier periods, is rarely significant.

The major diversification of birds probably took place in the Cretaceous, which lasted longer than the sum of all subsequent periods, and it must have started early in that period because fragmentary material of foot-propelled divers (*Enaliornis*) and of an early relative of the flamingos (*Gallornis*) are known from Lower Cretaceous deposits of Europe. Upper Cretaceous deposits have yielded, besides *Hesperornis* and *Ichthyornis* and their relatives, diving birds similar to *Enaliornis* (*Lonchodytes*), other early flamingo-like birds, and species in the same suborders as gannets, ibises, rails, and shorebirds.

Paleocene deposits have yielded the earliest known loons, cormorants, New World vultures, and gulls. In addition, the large, flightless predatory birds culminating in *Diatryma* first made their appearance during this period. From the far richer Eocene deposits have come the earliest known rheas, penguins, albatrosses, tropic birds, anhingas, true flamingos, herons, storks, secretary birds, hawks, curassows, cranes, bustards, avocets, auks, sand grouse, cuckoos, owls, swifts, trogons, rollers, hornbills, and songbirds. Almost certainly all living orders and most living families of birds were in existence by the end of the Eocene period. One of the most interesting finds from this period was fossils of *Neocathartes,* a long-legged bird allied to the New World vultures. There are several anatomical similarities between this group of vultures and the storks, and the existence of this fossil lends support to the idea that the storks and New World vultures are more closely related to each other than each family is to the birds with which it is usually grouped.

Important Oligocene fossils include the earliest phororhacoids, one of the few groups of fossil birds that is known from enough material from over a long enough time span to show evolutionary trends, in this case, both in size and in bill form.

Fossils of Miocene birds are numerous. Several early groups of pelecaniform, cranelike, and flamingo-like birds are known last from this period, and the first of the Mancallidae, superficially penguin-like auks, appeared. Otherwise, the avifauna was essentially modern.

By Pliocene times, most modern genera were probably in existence. The Mancallidae continued on the California coast at least until the middle of the period.

The appearance and extinction of large birds as well as mammals was a feature of the Pleistocene. Perhaps most notable were the teratorns, "super condors," which were found in North America. These included *Teratornis incredibilis,* the largest known flying bird.

CLASSIFICATION

Distinguishing taxonomic features. In classifying birds, most systematists rely primarily on structural characters. Plumage characters include the number of remiges and rectrices; the presence or absence of down on the feather tracts, on apteria, and on the oil gland; and the presence or absence of an aftershaft. Characteristics of the bill and feet are useful, as is the arrangement of bones in the palate and around the nostrils. The presence or absence of certain thigh muscles and the arrangement of the carotid arteries, the syrinx, and the deep flexor tendons of the toes are employed, as is the condition of the young when hatched. Advances in the study of protein structure and of chromosomes have provided new means of determining taxonomic relationships.

Annotated classification. This classification is based primarily on that of the American ornithologist Alexander Wetmore but includes the ideas of a number of other authorities. It is unlikely that most avian systematists would agree on all aspects of one arrangement, but the one presented below will satisfy many. The dagger (†) indicates extinct groups, known only from fossils.

CLASS AVES

Vertebrate (backboned) animals primarily adapted for flight with feathers. Warm-blooded, 4-chambered heart, left systemic arch lost. Lower jaw articulates with cranium via the quadrate; teeth absent in living forms. Reproduction by hard-shelled eggs, nearly always incubated by one or both parents. About 8,600 living species.

†Subclass Archaeornithes

†Order Archaeopterygiformes (Archaeopteryx). Upper Jurassic; Europe. Teeth set in sockets; long, unfused caudal vertebrae, each bearing a pair of rectrices; keelless sternum; functional claws on digits of hand. Gliding birds, about 50 cm long.

Subclass Neornithes

†Superorder Odontognathae

†Order Hesperornithiformes (Hesperornis, Baptornis). Upper Cretaceous; North and South America. Teeth set in groove in jaws. Flightless, foot-propelled diving birds, 1 to 2 m long.

Superorder Neognathae

Order Tinamiformes (tinamous). Upper Pliocene to present; Central and South America. Superficially quaillike or grouselike ground-dwelling birds with flat, elongated, and rather weak bills and very small tails. Size 15–50 cm.

Order Rheiformes (rheas). Lower Eocene to present; South America. Ostrich-like cursorial birds with very small tails and no aftershaft on the feathers. Sexes alike. Length 90–130 cm.

Order Struthioniformes (ostrich). Lower Pliocene to present (the Eocene *Eleutherornis* may belong here); southwestern Asia and Africa (fossils from southern Europe and southeastern Asia). 2-toed (3rd and 4th) running birds. Males black and white, females brown. Aftershafts, down, and filoplumes absent. Largest living bird; length to 180 cm, height 260 cm, weight 136 kg, egg 1.6 kg.

Order Casuariiformes (emus, cassowaries). Pleistocene to present; Australia, New Guinea, adjacent islands. Very large, cursorial (running) birds. Sexes alike, brown (emus) or blackish with brightly coloured wattles and skin on head (cassowaries). Aftershaft very large. Length 130–190 cm.

†Order Aepyornithiformes (elephant birds). Pleistocene; Madagascar (upper Eocene and lower Oligocene fossils from Egypt have been placed here). Very large and graviportal (heavy bodied); height to 3 m; egg weight estimated at 10 kg.

Order Dinornithiformes (moas, kiwis). Upper Miocene or lower Pliocene to present; New Zealand. Very large (to 3 m tall) and graviportal birds (moas) or smaller (length 30–80 cm); almost wingless, nocturnal, probing birds (kiwis).

Order Podicipediformes (grebes). Lower Miocene to present; worldwide. Foot-propelled diving birds with lobed toes, minute tails, and silky plumage. Length 21–66 cm.

Order Procellariiformes (albatrosses, shearwaters, fulmars, prions, petrels). Middle Eocene to present; all oceans, but most numerous in Southern Hemisphere. Web-footed marine birds with tubular nostrils; rhamphotheca divided into plates; possess a musky smell. Most have narrow wings and stiff, gliding flight. Length 14–135 cm.

Order Sphenisciformes (penguins). Upper Eocene to present; oceans of Southern Hemisphere. Wings flipper-like, for propulsion underwater; webbed feet short and stout; stance upright. Feathers short and dense, molted in patches. Length 40–120 cm (fossil forms to 180 cm).

Order Pelecaniformes (pelicans, boobies, tropic birds, cormorants, frigate birds). Paleocene to present; worldwide. Water birds with all 4 toes webbed; bill hooked or straight and sharply pointed. Length 50–180 cm.

Order Anseriformes (screamers, waterfowl). Middle Eocene to present. Web-footed birds with broad bills containing fine plates or lamellae (waterfowl); or large-footed marsh birds with chicken-like bills (screamers). Length 29–160 cm.

Order Phoenicopteriformes (flamingos). Cretaceous to present; discontinuously distributed in warm regions except Australasia. More varied and widely distributed as fossils. Web-footed birds with long legs, long necks, bent bills with lamellae, and much pink or red in the plumage. Share characters with both Anseriformes and Ciconiiformes, but evidently closer to the latter, with which they are sometimes grouped. Length 91–122 cm (some fossil forms smaller).

Order Ciconiiformes (herons, storks, ibises, spoonbills). Upper Cretaceous to present; worldwide except in extreme north. Long-legged wading birds with long bills; feet not webbed. Although usually grouped together, herons and storks may prove to belong to different orders. Length 28–152 cm.

Order Falconiformes (diurnal birds of prey). Upper Paleocene to present; worldwide. Diurnal raptors with hooked beaks, long talons, and short (hawks, falcons) or very long (secretary bird) legs or carrion-eating birds with weaker claws and tearing bills (vultures, condors). Length 15–150 cm (some fossil forms larger).

Order Galliformes (grouse, pheasants, quail, turkeys). Middle Eocene to present; nearly worldwide, except southern South America. Terrestrial or arboreal chicken-like birds; strong, scratching feet; short, rounded wings; feathers with long aftershafts. Length 13–198 cm.

Order Gruiformes (cranes, rails, coots, cariamas, bustards). Upper Cretaceous to present; worldwide. Diverse group, ranging from small quail-like hemipodes to large long-legged cranes, marsh-inhabiting rails, swimming coots and fin-foots, and cursorial bustards. The Tertiary phororhacoids belong here, as may the very large *Diatryma* and its relatives. Length 11–152 cm (fossils to 200 cm tall).

†*Order Ichthyornithiformes (Ichthyornis, Apatornis).* Upper Cretaceous; North America. Superficially gull- or ternlike water birds of uncertain affinities. Length approximately 21 to 26 cm (estimated from reconstruction of fossils).

Order Charadriiformes (plovers, sandpipers, gulls, terns, auks). Upper Cretaceous to present; worldwide. 3 basic body plans: Suborder Charadrii—waders (shorebirds), usually feeding on small animals in mud or water, bill variable but often long and used for probing; Lari—web-footed, dense-plumaged water birds feeding by plunging into water for fish, robbing other birds, or scavenging; Alcae—dense-plumaged, web-footed, marine, wing-propelled divers, feeding on fish or invertebrates. Length 13–76 cm.

Order Gaviiformes (loons or divers). Upper Paleocene to present; Holarctic. Foot-propelled diving birds with webbed feet and pointed bills. Cnemial crest an extension of the tibia. Length 66–95 cm.

Order Columbiformes (sand grouse, pigeons, doves, dodoes). Upper Eocene or lower Oligocene to present; worldwide except in extreme north. Fast-flying birds with pointed wings and weak bills; feed on seeds and fruit. Length 15–84 cm.

Order Psittaciformes (parrots, lories, cockatoos). Upper Oligocene to present; throughout tropics, with some temperate-zone species. Often brightly coloured. Strong-flying, seed-, fruit-, or nectar-eating birds with very stout, hooked bills and zygodactyl feet (*i.e.,* outer toe reversed). Length 9.5–99 cm.

Order Cuculiformes (turacos, cuckoos, roadrunners). Upper Eocene or lower Oligocene to present; worldwide except in extreme north. Long-tailed birds with zygodactyl or semizygodactyl feet. Feed on both fruits and small animals. Most arboreal, a few terrestrial. Some brood parasites. Length 16–70 cm.

Order Strigiformes (owls). Eocene to present; worldwide. Nocturnal raptorial birds with hooked beaks, strong talons, and soft plumage. Length 13–69 cm.

Order Caprimulgiformes (nightjars, frogmouths, oilbird). Pliocene to present; worldwide except in extreme north. Concealingly coloured, soft-plumaged, nocturnal birds with weak feet and very large mouths. Most feed on insects caught in flight. Length 19–53 cm.

Order Apodiformes (swifts, hummingbirds). Upper Eocene or lower Oligocene to present; worldwide except in extreme north; hummingbirds limited to New World. Rapid-flying birds that feed on the wing on insects and nectar. "Hand" and primaries constitute a relatively great proportion of the wing; feet weak. Length 6.3 to 23 cm.

Order Coliiformes (colies or mousebirds). Unknown as fossils. Africa south of the Sahara. Soft-plumaged birds with long, pointed tails and all 4 toes directed forward. Food largely vegetable, some insects. Length 29 to 36 cm.

Order Trogoniformes (trogons). Pantropical, except Australasia; upper Eocene or lower Oligocene to present. Extremely soft-plumaged arboreal birds; underparts yellow to red, head and neck often iridescent, tail long, black and white. Feet weak; 1st and 2nd toes directed backward. Food insects and small fruit. Length 23 to 34 cm.

Order Coraciiformes (kingfishers and allies). Eocene to present; worldwide except in extreme north. A heterogeneous group of hole-nesting birds. Many with long, pointed bills and blue or green in plumage. All have 2nd and 3rd or 3rd and 4th toes joined at base. Food largely animal, except hornbills, which eat much fruit. Length 9 to 160 cm.

Order Piciformes (woodpeckers, barbets, honey guides, toucans). Upper Oligocene (possibly upper Eocene) to present. Zygodactyl (rarely 3-toed) hole-nesting birds. Food insects and fruit. Woodpeckers are modified for climbing. Honey guides are brood parasites. Length 9 to 61 cm.

Order Passeriformes (perching birds). Upper Eocene to present; worldwide. The large complex assemblage of perching birds, containing more than half of the known species of birds. Bill, plumage, and habits highly varied. Length 7.5 to 102 cm.

Critical appraisal. It has frequently been stated that birds are one of the best known of animal groups. This is true, in the sense that most of the living species and subspecies in the world have probably been described; but because of inadequacies in the fossil record and repeated cases of convergent evolution within the group, our knowledge of the phylogenetic relationships between orders, suborders, and families of birds is inferior to that of mammals and reptiles.

The taxonomic positions of several bird groups remain open to question. The hoatzin, included above in the Galliformes, is often given its own order, Opisthocomiformes. The turacos, here included in the Cuculiformes, are considered by many authors to warrant separation as Musophagiformes. *Diatryma* and several related genera of extinct flightless predators are often placed in a distinct order, Diatrymiformes, near Gruiformes. The flamingos, which constitute the order Phoenicopteriformes above, are placed in the Ciconiiformes in many classifications. The sand grouse, family Pteroclidae, are believed by some to be more closely related to the shorebirds (order Charadriiformes) than to the pigeons (order Columbiformes), with which they are usually grouped.

Relationships among the ratites

One area particularly in need of study is the relationships among the various groups of ratites (ostriches, rheas, emus, moas, etc.). Formerly, some authorities argued that these birds and the penguins arose independently from cursorial reptiles, but it is now generally agreed that all of them passed through a flying stage in the course of their evolution. The ratite groups differ greatly in morphology and yet show remarkable similarities in palate and bill characters. The principal unanswered questions are how many different flightless lines evolved from flying ancestors and from how many different groups were the flying ancestors evolved. On zoogeographic grounds, it is likely that the isolated kiwi-moa, elephant bird, and emu-cassowary lines arose independently from each other and from ratites on the other continents. But the ostriches and rheas could be descended from a common flightless ancestor because of the known former land connections from Asia to North and South America.

Before organic evolution was understood and accepted, animals were grouped on the basis of general similarity. It is now known that many such groupings were unnatural from a phylogenetic standpoint but were instead the result of convergent evolution from different parental stocks. Examples are *Hesperornis,* loons, and grebes, and diving petrels and auks. It is likely that many more examples are not recognized or generally accepted. At least the following groups should be studied with this in mind: herons and storks, *Diatryma* and the phororhacoids, New World vultures and other falconiforms, sand grouse and pigeons, touracos and cuckoos, and swifts and hummingbirds. These examples are all from the ordinal or subordinal level; examples at lower levels would be far more numerous. (R.W.St.)

MAJOR BIRD ORDERS

The remainder of this article consists of a review of each of the major groups of birds, identified at the order level of biological classification and arranged in accordance with the annotated classification above.

There are some omissions, however. The extinct groups (designated in the annotated classification by the dagger [†]) are treated throughout this article under the heading *Evolution and paleontology.* Additional information on extinct birds and on avian evolution may be found in GEOCHRONOLOGY: *Fossil record.* Also omitted are the following orders: Rheiformes (rheas); Struthioniformes (ostriches); Dinornithiformes (moas and kiwis); Gaviiformes (loons, also called divers); Coliiformes (colies, also called mousebirds); and Trogoniformes (trogons). These groups are represented by *Micropædia* entries under their common rather than scientific names. Finally, the flamingos (Phoenicopteriformes) are treated in the section on Ciconiiformes.

Tinamiformes (tinamous)

The Tinamiformes, or tinamous, are a group of ground-dwelling, chicken-like birds of Central and South America. They have a superficial resemblance to partridges and quail (gallinaceous birds) but are placed in a distinct order related to the much larger rheas. Tinamous long have interested scientists because of many peculiar features of their skeleton and biology that link them to the large flightless birds or ratites: the ostriches, emus, cassowaries, and rheas.

Predation by man

Tinamous, considered by hunters to be among the finest game birds in terms of sport as well as palatability, are heavily hunted in many parts of South America. Market hunting has been curtailed by law; however, it is still practiced in some countries. Frozen tinamous from Argentina were formerly sold in the United States under the name South American quail. Although by the early 1980s only one species of tinamou was listed as endangered by the International Union for the Conservation of Nature, habitat destruction and heavy hunting have reduced a large number of populations.

GENERAL FEATURES

The 46 species of tinamous range in size from that of a small quail, about 15 centimetres (six inches) long and 150 grams (five ounces) in weight, to that of a large grouse, about 50 centimetres long and two kilograms (four pounds) in weight. Tinamous are rather uniform in body proportions and stance, resembling guinea fowl (Numididae). The head is small and the bill medium sized, relatively thin, and slightly downcurved. The long, slender neck is clothed in short feathers. The body is quite heavy, with a high rump outline from the enormous development of rump feathers, which generally hide the extremely short or even rudimentary tail. The short, rounded wings are inconspicuous on the standing bird, and the primary flight feathers are hidden by the full plumage of the flanks. The bare legs are typically rather thick and of medium length. There are three short front toes, with the hind toe either elevated or absent.

Protective coloration in spotted or barred patterns

The sexes are alike, except that the female is generally slightly heavier and has brighter coloration. The plumage coloration is highly concealing, in spotted or barred patterns of brown, gray, rufous, or tan. The variation in coloration is dependent upon the environment. The crested tinamous of the genus *Eudromia* have a long and slender crest that the bird directs forward when it is excited. The colour of the legs or of the bill is vivid and diagnostic in several species, as for example in the yellow-legged tinamous (*Crypturellus noctivagus zabele*).

NATURAL HISTORY

Locomotion. Highly adapted for ground dwelling, tinamous normally walk rapidly (especially the savanna species) and can run with amazing swiftness. If forced into extended running, however, as when chased by men or dogs, they tire quickly and are likely to stumble and fall. They are best able to escape notice by standing motionless with the neck extended or by quietly slipping away, making use of all available cover. Some species of tinamous may crouch or even feign death. They rise in flight only when almost stepped upon. Small tinamous that live in open terrain sometimes hide in animal holes, such as the burrows of armadillos.

The flight of tinamous is clumsy but swift and accompanied by an easily audible rumbling or whistling noise produced by the stiff, curved primaries. The elegant crested tinamou (*Eudromia elegans*) of the open tableland of Argentina alternates periods of flapping with short glides. When flushed, forest species sometimes collide with branches and tree trunks and may injure themselves. If forced to make several flights in short succession, tinamous soon become exhausted, apparently because of a low circulation rate, related in turn to the surprisingly small size of the heart and lungs. The flight muscles are well developed, but the circulatory system seems to be insufficient for sustained activity.

Unlike the gallinaceous birds, tinamous sleep on the ground at night. Exceptions are members of the genus *Tinamus,* which roost in trees, choosing horizontal branches or tangled lianas and perching on their rasplike feet without using the toes.

Vocalizations. The voices of tinamous are among the strongest and most pleasant of any in the neotropics. They consist of loud, stereotyped, but melodious whistles, varying from the long and astonishingly songlike sequence of the brown tinamou (*Crypturellus obsoletus*)—astonishing because most relatives of the tinamous do not produce elaborate vocalizations—to the monosyllabic call of the cinereous tinamou (*C. cinereus*). The calls of the male and female are similar but discernibly different to the human ear. The female solitary tinamou (*Tinamus solitarius*) has a special call given during the time before egg laying, and another call is uttered by both sexes of this species after perching at dusk.

Habitat selection and food habits. Collectively, tinamous are adapted to a wide variety of environments, including dense woodland, thickets, open woodland, savanna, and even to the bunchgrass-covered plateaus of the high Andes, where they occupy the ecological niche generally held by the grouse in other parts of the world. In some forest regions as many as five species of tinamous have been found to coexist, inhabiting slightly different types of plant communities. The grassland tinamous do not occur north of the Amazon River; there the tinamou niche is occupied by the crested bobwhite (*Colinus cristatus*), a species of quail.

Seasonal variation in feeding

The food taken by tinamous varies with the season and habitat. In summer the red-winged tinamou (*Rhynchotus rufescens*), for example, eats mainly animal material—largely insects, but its mouth is large enough to swallow mice. In the stomach of one bird 707 termites were counted. In winter the red-winged tinamou shifts over to vegetable food. It occasionally becomes a pest in agricultural areas, using its strong bill to dig up the roots of manioc, or cassava (*Manihot esculenta*). The small tinamous of the genus *Nothura* feeds primarily on seeds, but the spotted tinamou (*Nothura maculosa*) occasionally eats ticks in pastures. The forest-inhabiting solitary tinamou generally prefers small fruits and berries, collected on the ground. It might, however, also devour a frog when it finds one. The members of the genus *Nothoprocta* are considered beneficial to agriculture because of their large consumption of insect pests. Young tinamous of all species are more dependent upon insects than are the adults. Unlike the gallinaceous birds, tinamous do not scratch for food, as is evident by their weak toes and short nails; instead, they either turn over leaves and other debris with the bill or dig with it.

Reproduction. Courtship behaviour has been described for only a few species of tinamous. Certain species have well-defined breeding periods, and others breed throughout the year. Investigators have reported that courting birds raise the thickly feathered rump and display the brightly coloured undertail coverts. A similar display has been observed in a frightened *Crypturellus:* the bird presses the breast to the ground, raises the rump, spreads the terminal feathers like a fan, and exhibits the sharply marked undercoverts. Courting birds also have been observed to chase each other around on the ground.

Multiple mating is the rule among tinamous, although a few species such as the ornate tinamou (*Nothoprocta ornata*) maintain stable pairs. All forms of polygamy exist, the conditions varying between and even within species. Many species have uneven sex ratios; preponderance of males seems to be more frequent, the ratio reaching four to one in the variegated tinamou (*Crypturellus variegatus*). The ratio in the ornate tinamou is about one to one.

Parental behaviour by the male

The nest, a shallow depression in the ground, is constructed and defended by the male. The eggs are among the most beautiful of all bird eggs, always monochromatic and highly glazed. The colours include light chocolate brown, near black, purple, dark bluish green, light yellowish green, and gray when laid, but the shell pigments fade when exposed to light. One hen, or more, places her eggs in the nest of a male; when several females provide the eggs, the clutch may become quite large, eight to 16 eggs. Incubation, which lasts 17 to 21 days, is done entirely by the male, who broods and guides the chicks for several weeks after hatching. The chicks, blotched and streaked like the young of the rheas, are able to run as soon as they are hatched. When frightened, they squat and freeze, becoming almost invisible.

PALEONTOLOGY AND CLASSIFICATION

There is no doubt that the tinamous represent one of the oldest stocks of birds on the South American continent. To date, three genera of fossil tinamous, of one species each, have been described from a single deposit from the Miocene of Argentina (about 10,000,000 years ago). The majority of other fossil tinamous, mostly representing species still extant, has been discovered at scattered sites from the upper Pleistocene (less than 1,000,000 years ago) of South America.

Many authors have noted anatomical and biological resemblances between the tinamous and rheas, or nandus (Rheidae). The structure of the bony palate, an important feature in the taxonomy of ratite birds, quite clearly links the two groups, but most authorities prefer to maintain them as separate orders, Rheiformes and Tinamiformes, each with a single family, respectively, Rheidae and Tinamidae.

(H.S.)

Casuariiformes (emus, cassowaries)

The order Casuariiformes includes two families of large flightless birds: Dromaiidae for the emu (*Dromaius novaehollandiae*), found only in Australia; and Casuariidae for three species of cassowaries (*Casuarius*), restricted to northern Australia, New Guinea, and nearby islands. Of the two groups, the emu is far better known, both biologically and popularly, being exhibited in zoos around the world.

The emu was first identified by European explorers in 1788, more than a century after the first cassowaries had been seen by Europeans; in 1697 a Dutch navigator, Willem de Vlamingh, had seen the emu's footprints in western Australia and had attributed them to a "Kasuarius." Cassowaries first became known to Europeans in the 17th century—there was a published reference to a "Casoaris" in 1658—when the Portuguese and Dutch colonized the East Indies. Both names, emu and cassowary, were originally applied to cassowaries; the emu was known as the New Holland or New South Wales cassowary until the early 19th century, when the name emu was gradually transferred to it. By the late 19th century about 11 species of cassowaries were recognized, but greater understanding of variation within species has reduced the number to three: the double-wattled, or Australian, cassowary (*Casuarius casuarius*), the single-wattled, or one-wattled, cassowary (*C. unappendiculatus*), and the dwarf, or Bennett's, cassowary (*C. bennetti*).

IMPORTANCE TO MAN

The emu and cassowaries are important foods for the indigenous peoples of both New Guinea and Australia; the flesh of the thigh muscles somewhat resembles beef. In New Guinea cassowaries are captured as chicks and held in enclosures until they are large enough to eat. The feathers are frequently used for personal adornment. In Australia cassowary feathers formerly were used for the notorious *kurdaitcha* shoes, or "shoes of silence," worn by Aboriginal executioners on nocturnal missions of tribal vengeance or punishment. During World War I members of the Australian cavalry regiments wore tufts of emu plumes on their slouch hats.

Since the time of European settlement in southeastern Australia, the emu has become much less abundant, and three island forms have been exterminated by hunting. In Western Australia, however, the bird has remained common and has become a pest in the wheat-farming areas, breaking fences and trampling and eating crops. In 1932 members of an army machine-gun unit were employed in a campaign against a concentration of emus, estimated to be about 20,000, in the vicinity of the wheat-belt centre of Campion.

The Emu War in Australia

The outcome of this bizarre Emu War, as it was called, has been summarized by the ornithologist D.L. Serventy as follows:

> . . . the machine-gunners' dreams of point blank fire into serried masses of Emus were soon dissipated. The Emu command had evidently ordered guerrilla tactics, and its unwieldy army soon split up into innumerable small units that made use of the military equipment uneconomic. A crestfallen field force therefore withdrew from the combat area after about a month.

Yearly kills of emus in Western Australia for bounty payments vary from about 5,000 to nearly 40,000 birds. The emus, however, remain plentiful, and small parties may be seen within 15 or 20 kilometres (9 or 12 miles) of Perth, the capital of the state.

NATURAL HISTORY

Reproduction. Throughout its climatically varied range the emu is a winter breeder; egg laying begins at the end of April. The nest is a flattened bed of bark, grasses, and leaves near a tree or bush and is so situated that the sitting bird (always the male) has a good view of its surroundings. Despite their size, nests are extremely difficult to find. The large green eggs, with granulated shells, average about 130 millimetres (slightly over five inches) in length and 87 millimetres (3.4 inches) in width and weigh 450 to 800 grams (16 to 28 ounces). They are laid at intervals of about four days, and the male starts incubation when the hen has laid five to nine. The normal clutch size is eight, nine, or 10 eggs, with large clutches up to 16. In exceptionally good seasons as many as 20 eggs may be laid, in poor seasons as few as four or five. Incubation varies from about 58 to 61 days, during which the male seldom leaves the nest, even to feed.

Eggs

The newly hatched chicks, which are concealingly streaked

Drawing by R. Keane

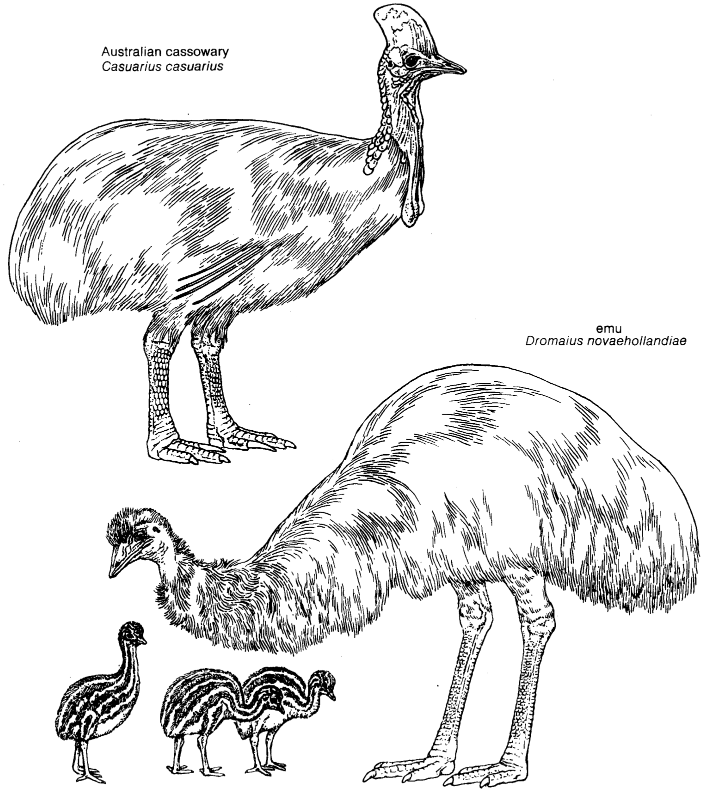


Figure 9: Emu and cassowary.

with black, are brooded by the male for two or three days before they begin to move around in his company. Normally two years are required to reach maturity. In the nonbreeding season the groups coalesce into flocks, which may undergo local movements or more extensive migrations. Marking with leg bands has proved that individual emus may travel hundreds of kilometres.

Far less precise data are available on the annual cycle of the cassowaries. According to several observers, the eggs are laid in winter and are incubated by the male for 49–52 days. The eggs are paler green than those of the emu and there are fewer in the clutch, which varies from three to eight. Maturity may not be reached until the second or third year. Cassowaries are not gregarious in the nonbreeding season but live as pairs or family parties.

Ecology. Cassowaries are more pugnacious and aggressive than the emu. They are essentially creatures of the tropical rain forests, where they are extremely difficult to observe. Their diet consists mainly of fruit. Emus, on the other hand, live under a wide variety of environmental conditions, ranging from dense temperate forest to open plains and arid scrub country. They frequently exhibit a curious inquisitiveness toward the activities of man and may be lured to close range by various devices, as for example the waving of objects in the air, a habit that was utilized by the Australian Aborigines in hunting them. Emus are basically vegetarian, subsisting on fruits, seeds, and vegetation of all kinds. They will, nevertheless, also feed on insects, including caterpillars and grasshoppers, when they occur in great numbers.

The emu and cassowaries swim well and are fast runners. The emu is capable of a stride of nearly three metres (10 feet) and can keep up with an automobile travelling as fast as 40 to 50 kilometres (approximately 25 to 30 miles) per hour for some distance.

Vocalizations. Both sexes of the emu are reported to utter resonant drumming notes and guttural grunting calls, under varying conditions. Less is known about the calls of cassowaries, but scattered accounts indicate that they are similar to those of the emu.

FORM AND FUNCTION

Size and weight of emus

Next to the ostrich (*Struthio camelus*), the emu is the largest and heaviest of living birds; large individuals stand between 1.5 and 1.8 metres (five and six feet) high and weigh up to 55 kilograms (120 pounds), although most are well below that weight. Females are slightly larger, averaging about 40 kilograms (90 pounds) while the males average about 36 kilograms (80 pounds). The sexes are alike in colour; the plumage is sombre brown, or brownish black, and the naked skin on the head and neck is blue. All three species of cassowaries are black and, except for the dwarf cassowary, have fleshy pendulous wattles of red, orange, or yellow on the head and neck. The head is crowned with a horny casque or helmet, believed by some authorities to protect the bird from injury by branches during rapid movements in the forest. The plumage of casuariiforms is loose and hairlike due to the lack of barbules, the secondary branches that interlock to form the flat vane in the feathers of most birds. There are no feathers differentiated as tail feathers, but cassowaries have five quills of the wing modified into hollow, unbranched spines.

As with many running birds, casuariiforms have only three toes, the hindtoe having been lost. The inner toe of cassowaries is armed with an elongated, dagger-like claw, making the foot a formidable weapon in kicking.

The casuariiform skeleton is similar to that of the other large flightless birds (ratites) in the reduction of the wing elements and of the keel on the sternum (breastbone) and in the enlargement of the leg elements. Vestigial clavicles (collarbones) remain in the shoulder girdle, but the humerus (upper "arm" bone) is much reduced, being shorter than the combined forewing and manus ("hand"). The manus has a single digit, now believed to be the third, which bears a long claw.

The pattern of bones in the palate, an important diagnostic feature in the taxonomy of ratites, is of the palaeognathous or dromaeognathous type (common to all ratites), in which the vomer bones of the skull extend back to separate the palatines. The casuariiforms have the simplest form of this palate type, with large vomers and short palatines.

EVOLUTION AND CLASSIFICATION

Paleontology. No fossil remains of casuariiform birds have been found outside of the Australian region, and most of those recorded are from the Pleistocene (more than 10,000 years ago), with one doubtfully from the Pliocene (about 4,000,000 years ago). The latter, although definitely a member of the emu assemblage, showed features linking it with the cassowaries. The absence of fossil material notwithstanding, the order must have had a long evolution during the Tertiary Period (from 65,000,000 to 2,500,000 years ago). The relationship of the casuariiforms with the other orders of large flightless birds—the rheas (Rheiformes), ostriches (Struthioniformes), kiwis and moas (Dinornithiformes), and elephant birds (Aepyornithiformes), formerly grouped in a distinct subclass of birds, the Ratitae—is still rather uncertain, but Casuariiformes appears to be the most primitive of the living orders. The upper Pleistocene of Australia has produced several fossil species of emus assigned to the modern genus *Dromaius* and dating from as much as 100,000 years ago. Other Pleistocene finds from Australia have been placed in a separate family, Dromornithidae. These include *Dromornis,* known only from leg bones, and the gigantic *Genyornis newtoni,* almost the entire skeleton of which is known. The only known fossil cassowary is a species from Pleistocene cave deposits in the Wellington Valley, New South Wales, Australia, which resembles closely, and is probably identical with, the dwarf cassowary (*Casuarius bennetti*), now restricted to New Guinea and nearby islands. The distribution of the Casuariidae has evidently shrunk in recent millennia; the only species occurring on the Australian mainland is the double-wattled cassowary (*C. casuarius*), which is restricted to New Guinea and the Cape York Peninsula of northern Queensland.

Relationship to other flightless birds

Classification. The following classification of the order is universally accepted.

ORDER CASUARIIFORMES

Large graviportal (ponderous) flightless birds with 3 toes. Other physical features described in *Form and Function,* above. Monogamous. Eggs green in colour.

Family Casuariidae (cassowaries)

Beak laterally compressed. Prominent horny casque on mesethmoid, nasal, and frontal bones of the skull. Coracoid bone of the shoulder much shorter than in emu. Bony union (symphysis) between various posterior elements of the pelvis may be present or absent. Femur not pneumatized by air sac system. Cervical vertebrae 18 or 19. Wing with 5 quills modified as long hollow spines. Pendulous coloured wattles on throat and neck (except in dwarf cassowary). Three species.

Family Dromaiidae (alternatively Dromiceiidae; emu)

Beak flattened dorsoventrally. Head not casqued. Pubic symphysis lacking. Femur pneumatic. Trachea with an aperture in front, leading to an inflatable neck sac. Pendulous wattles absent. Fossil record limited to the Pleistocene of Australia and nearby islands. One Recent species.

Family Dromornithidae

Fossil only; Pleistocene of Australia; 2 genera, 2 species.

(D.L.Se.)

Podicipediformes (grebes)

The order Podicipediformes includes only the grebes, foot-propelled diving birds best known for the striking courtship displays of some species and for the silky plumage of the underparts, which formerly was much used in millinery. The speed with which grebes can submerge has earned them such names as water-witch and helldiver, while the position of the feet near the tail is responsible for the early English name arsefoot, from which the family name, Podicipedidae, was derived.

Adult grebes range in weight from less than 150 grams (five ounces) to more than 1.4 kilograms (approximately three pounds) and in total length from 21 to 73 centimetres (eight to 29 inches). They vary principally in bill shape and ornamentation of the head. The group is found on all of the continents and on many island groups as well.

However, it is best represented in temperate regions. Seven species each are found in North and South America, five in Eurasia, and three each in Africa and Australia. The species range from conspicuous and gregarious to solitary and skulking.

NATURAL HISTORY

Courtship displays

Vocalization

Mating behaviour. Both parents share in nest building, incubation, and care of the young. The long pair bond that makes this possible is formed and strengthened by elaborate courtship displays, including ritualized preening, head shaking, diving, weed carrying, and rapid water treading with the body in a nearly vertical position. These displays may be combined into complex ceremonies such as the discovery ceremony of the great crested, rednecked, horned, eared, and related species or the rushing display of the western grebe. In nearly all courtship ceremonies the roles of the sexes are interchangeable. The same is true of the precopulatory displays, and reverse mounting has been reported for all species that have been thoroughly studied. Courtship feeding is known only in the western grebe (*Aechmophorus occidentalis*), the male of which feeds the female. Grebe vocalizations include advertising calls, copulation trills, "conversational" notes, and duetting trills. In the courtship of more secretive species, such as the pied-billed grebe (*Podilymbus podiceps*) and the dabchicks (a name given to several of the smaller grebes), vocalizations are relatively more important than visual displays.

Nesting. Following pair formation, grebes build one or more floating platforms of aquatic vegetation. On these mating takes place and the three to seven unmarked chalky, white eggs are laid. On leaving the nest, adults often cover the eggs with the soggy nest material, and, by the end of the incubation period (usually three to four weeks), the eggs are strongly stained. In most cases, the nest is deserted after the young hatch, and the small young spend most of their time on the swimming parents' backs, which they reach by clambering up the adult's foot. On hatching, the downy young are marked with bold longitudinal stripes, though such stripes are rarely visible on the gray young of the western grebe. Remnants of the head stripes are carried over into the juvenile plumage. In addition to the stripes, the young have either one or more patches of rufous down or a bare spot on the crown. The skin of the latter changes in colour from pink to red when the young become excited.

Feather eating

Habitat selection and food habits. Grebes breed on still or slow-moving bodies of fresh water. The dabchicks and the pied-billed grebe are most numerous on small waters with much emergent vegetation, the western grebe on large bodies of open water. Some species winter on sheltered coastal waters. Grebes' diets consist almost entirely of animal matter. The western grebe feeds largely on fish, the eared grebe (*Podiceps nigricollis*) on small invertebrates, and the pied-billed grebe takes many heavy-bodied crustaceans. Other species, such as the horned and red-necked grebes (*Podiceps auritus* and *P. grisegena*), have more varied diets. Grebes are noted for swallowing feathers, which form a plug in the pyloric pocket of the stomach and effectively filter material passing to the intestine. Adults feed feathers to the young, establishing this plug shortly after hatching. Grebes feeding on invertebrates containing much chitin regurgitate pellets rather frequently; fish eaters tend to maintain masses of feathers in the main portion of the stomach, presumably to hold fish bones until they are digested.

FORM AND FUNCTION

Grebes can be distinguished by their lobed toes with flat nails, greatly reduced tail feathers, and satiny breast feathers. The last is due to the relatively loose structure of the feathers on this part of the body and to spirally coiled barbules (the secondary branches of the feathers), which lie parallel to the barbs (the primary branches). As with most diving birds, the conspicuous features that distinguish species from one another are found on the head and figure prominently in courtship.

Western grebe (*Aechmophorus occidentalis*). The young are carried upon the parents' backs for protection and warmth.
Willis Peterson

Seasonal plumage

In winter plumage most grebes are brown, black, or gray above and white or light brown below. In summer plumage, rufous, buff, black, or white markings or elongated plumes are found, chiefly on the head and neck. In this plumage, the males tend to be more brightly coloured and to have somewhat longer plumes than the females. The males also tend to be larger and larger billed than the females.

Grebes' wings are short and the wing bones small in diameter. Some species nesting in cold regions are migratory. Major flights are usually made at night. One species, the short-winged grebe (*Rollandia micropterum*) of Lake Titicaca, between Peru and Bolivia, is flightless; two others, Taczanowski's grebe (*Podiceps taczanowskii*) of Lake Junín, Peru, and the giant pied-billed grebe *Podilymbus gigas*) of Lake Atitlán, Guatemala, have somewhat reduced wings and fly little, if at all.

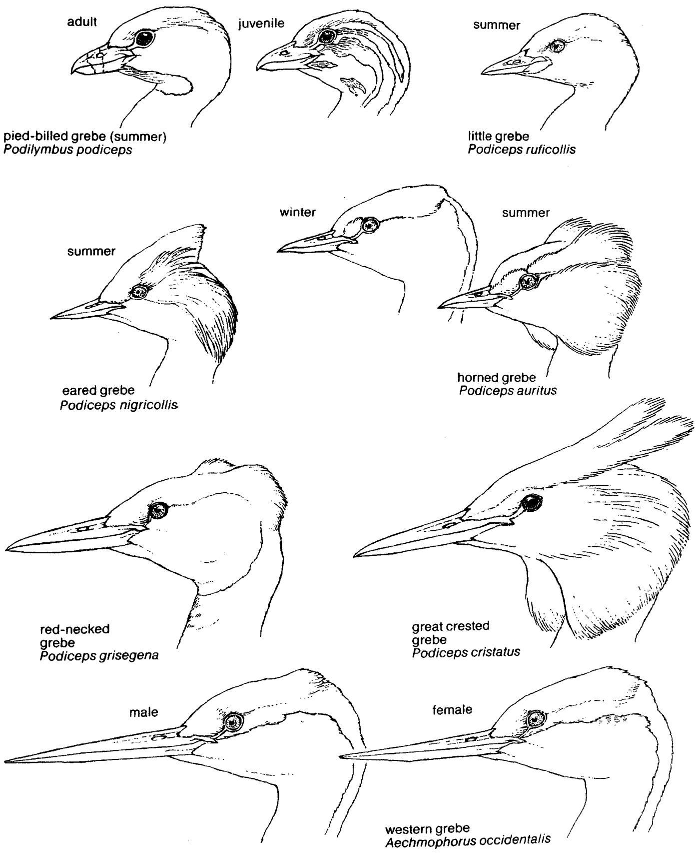

Drawing by R. Keane

Figure 10: Heads of typical grebes.

While on the surface, grebes usually swim with alternate strokes of the feet, whereas under water, the feet move simultaneously. Between recovery and power strokes, the feet are turned 90° so that the broad side can be used for propulsion; just before the recovery stroke, the feet are rotated in the opposite direction. While resting, grebes keep one or both feet shipped under the wings, and the head is kept forward with the bill under the neck (the "pork-pie" posture).

CLASSIFICATION AND PALEONTOLOGY

Grebes comprise a single family, Podicipedidae, of the order Podicipediformes. There are approximately 18 species, usually placed in five genera: *Aechmophorus* (the western grebe), *Podiceps* (most species), *Podilymbus* (the pied-billed grebes), *Rollandia* (Rolland's and short-winged grebes), and *Tachybaptus* (dabchicks).

Distinguishing taxonomic features. The order and family are defined on the basis of structural features, especially the form of the nostrils, the absence of certain processes in the skull, and the absence of the ambiens muscle in the leg. The knee process consists of a large pyramidal patella (kneecap) and a large projection on the tibia (the main bone of the lower leg). The wing is diastataxic (*i.e.,* there is a space where the fifth secondary feather is located in many birds), with 12 primaries, seven of which are attached to the metacarpus (midhand). The oil gland is tufted and has two openings. Only the left carotid artery is present. The intestinal caeca (blind pouches) are small.

Grebes are highly specialized diving birds with no obvious relatives, living or fossil. Statements to the effect that they are "primitive" or "reptilian" are erroneously based on a superficial resemblance to the toothed Cretaceous Hesperornithiformes, themselves very specialized forms. The earliest known fossil grebe, *Podiceps oligoceanus* from lower Miocene deposits (about 25,000,000 years old) in Oregon, is essentially like modern species.

The names *Colymbus,* Colymbidae, and Colymbiformes were formerly applied to the grebes in North America and to the loons (*Gavia*) in Europe. (R.W.St.)

Procellariiformes (albatrosses, shearwaters, petrels)

The Procellariiformes are a distinct natural order of oceanic birds with approximately 87 living and 36 fossil species. Of diverse size and range, they are divided into four families: albatrosses, shearwaters, storm petrels, and diving petrels. All Procellariiformes are recognizable by their conspicuous tubular nostrils, which project upon the culmen (upper bill). This feature gives the order its alternative name, Tubinares, meaning "tube nosed." The feet of these birds are webbed, and the hind toe is either vestigial or missing. All species have a characteristic powerful musky odour, which is caused by the excretion of stomach oil. The oil can be used as a defensive discharge through the mouth when the bird becomes alarmed.

GENERAL FEATURES

Importance to man. The tube-nosed birds have been of considerable local economic importance as a source of protein food, feathers, and oil wherever humans have colonized or have been able to raid the coastal and oceanic islands where they breed. This situation has resulted in the partial or complete extermination locally of certain species. Humans, moreover, have been responsible for the introduction of various predators, including rats, pigs, and cats. In regions where bird populations have survived, people have continued to harvest the eggs, the plump young birds (at fledging time), or both. Many thousands of slender-billed, or short-tailed, shearwaters (*Puffinus tenuirostris*) are taken on the Bass Strait islands off Tasmania and sold fresh, salted, or deep-frozen as "muttonbirds." In all likelihood, the name muttonbird was derived from the use of the flesh as a supplement for mutton by the early settlers of New South Wales. The numbers of muttonbirds now harvested are regulated so as to preserve a substantial breeding stock.

"Mutton-birds"

In New Zealand the Maori people have harvested young *titi* (shearwaters of several species) from time immemorial, a right assured them in perpetuity by treaty with Queen Victoria. On the other side of the world, hundreds of Manx shearwaters (*Puffinus puffinus*) were formerly collected for food and as lobster bait on the Welsh islands of Skomer and Skokholm, which are now nature preserves estimated to contain about 200,000 Manx shearwaters and 2,000 storm petrels (*Hydrobates pelagicus*). On the Tristan da Cunha Islands in the South Atlantic, resident islanders harvest the eggs and squab (young) of a large, mixed seabird population, which includes more than 6,000,000 greater shearwaters (*Puffinus gravis*).

The harvesting of northern fulmar petrels (*Fulmarus glacialis*) is an ancient practice among peoples inhabiting the cool northern coasts where the birds choose to breed. In Iceland about 50,000 fulmars were taken annually between 1897 and 1925; however, the occurrence in 1939 of psittacosis (a virulent avian disease) among processors of the birds resulted in prohibition of the use of fulmars for food.

During the early 17th-century colonization of Bermuda, millions of cahows, or Bermuda petrels (*Pterodroma cahow*), were exterminated by the colonists. For nearly 300 years the species was believed extinct, but in 1951 a few pairs were discovered nesting on an offshore islet, where a remnant now survives under strict protection. The related black-capped petrel, or diablotin (*P. hasitata*), of the West Indies was also thought extinct (because of predation by humans, rats, and mongooses) until in 1961 a substantial population, estimated to number at least 4,000 birds, was found breeding in the inaccessible forested crags of Hispaniola.

Encounter with man

In the 18th and 19th centuries, huge numbers of albatrosses were taken for food (largely by whalers) and for the millinery trade. With the disappearance of sailing vessels, changes in fashions, and the establishment of many nesting grounds as sanctuaries, such predation has virtually disappeared, but albatrosses have not entirely escaped stress at the hands of man. On Sand Island in the Midway Atoll, Laysan albatrosses (*Diomedea immutabilis*) increased from a few pairs in 1900 to about 60,000 pairs in the early 1960s, the increase resulting from shelter provided by introduced vegetation. The use of the island by aircraft after 1935, air raids by the Japanese during World War II, and the loss of 30,000 birds humanely killed by the U.S. Navy (in a control program designed to reduce collisions between birds and aircraft) did not deter the albatrosses from this favoured nesting area. The control program was abandoned after the discovery that levelling

Drawing by R. Keane

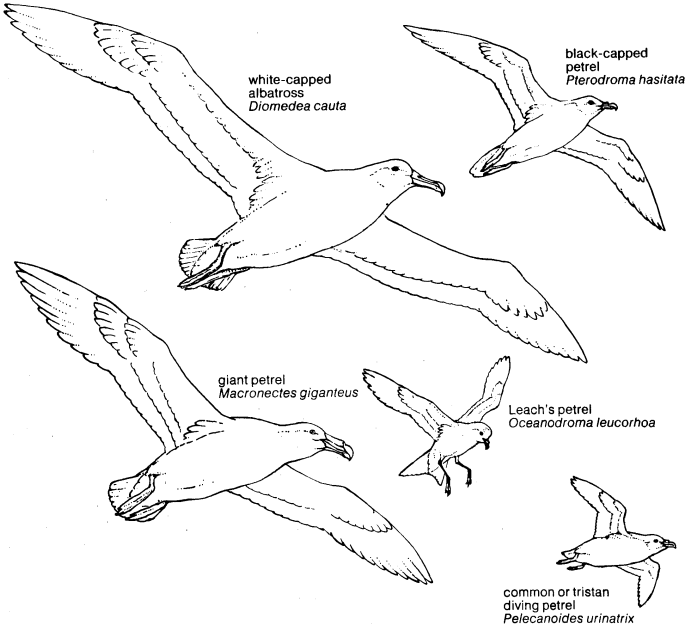


Figure 11: Representative procellariiform birds in flight.

certain sand dunes effected a 70 percent reduction in collisions by removing the updrafts near aircraft runways.

Distribution. The majority of procellariiforms breed in the Southern Hemisphere, but several species migrate thousands of miles north across the Equator to winter in the northern summer seas, where they molt, feed, and rest in preparation for the return home in the southern spring. Similarly, species that breed in the Northern Hemisphere also live in perpetual summer by migrating far south to winter in the southern summer. A number are less migratory and do not cross the Equator. Several species are almost sedentary, chiefly smaller petrels breeding in tropical and subtropical latitudes, and the subantarctic prions (*Pachyptila*) and diving petrels. All latitudes of the unfrozen oceans are thus occupied, but there are fewer living in the calm equatorial region, where there is little wind to lift their long wings and where the pelagic (open-ocean) crustacean food on which so many seabirds basically depend is scarce. The zone of upwelling water in the windy latitudes of the Antarctic convergence, between 40° and 60° south latitude, is richest in the shrimplike krill (*Euphausia* species), attracting surface-feeding tubinares and the diving penguins, prions, and diving petrels. Some feed along the edge of the ice of the Antarctic continent, and four tubinares actually breed on its shores: the Antarctic fulmar, the giant petrel, the snowy petrel, and the tiny but very numerous Wilson's petrel; the nesting burrows of the last may be blocked by snow for days during the protracted breeding season. The only tubinare nesting near the ice limits in the high Arctic is the fulmar, which reaches Franz Josef Land, Greenland, and the Arctic Circle north of the Aleutian Islands.

Of the albatrosses (family Diomedeidae), only the two Midway species and the short-tailed albatross (*Diomedea albatrus*) nest well north of the equatorial doldrums. The latter was brought close to extinction by plume hunters and by a volcanic eruption at its nesting island of Torishima. There were enough immature birds at sea at the time to allow a partial recovery, and some 60 individuals were counted there in 1969. Nine albatross species range the Southern Hemisphere, gliding on the eternal winds of the "Roaring Forties" (the region between 40° and 50° latitude) and moving north with the food-rich cold currents along the west coasts of South America, South Africa, Australia, and New Zealand. One species, the waved albatross (*Diomedea irrorata*), is unique in that it breeds only in the Galápagos Archipelago at the Equator, where probably not more than 3,000 pairs nest on Hood Island.

The family Procellariidae includes the larger petrels, such as the northern and southern fulmars (*Fulmarus glacialis* and *F. glacialoides*), the gadfly petrels (*Pterodroma*), several genera of shearwaters, and the prions or whalebirds. Several of the shearwaters and larger petrels breed in burrows far inland on mountain crags in the Andes, West Indies, Madeira, and New Zealand. The largest member of this family is the giant, or stinker, petrel (*Macronectes giganteus*), an albatross-like scavenger and circumpolar wanderer with a heavy beak and wing span of eight feet. Smallest are the prions, four species of small, stocky, little-studied birds, 22 to 30 centimetres (nine to 12 inches) long, with broad bills and a restricted cold-water range; they breed on sub-Antarctic islands, keeping much to the water, as do the diving petrels.

The storm-petrel family, Hydrobatidae, ranges both hemispheres but is strongest numerically in the Pacific, where *Halocyptena microsoma,* the least petrel of Baja California, rivals the Atlantic storm petrel as the smallest procellariiform. The word petrel ("little Peter") derives from a habit of the storm petrels of walking on the waves.

The diving petrels form a family (Pelecanoididae) and genus (*Pelecanoides*) with four species. They are small, rather sedentary, coast-dwelling birds confined to cool southern islands, including Tristan da Cunha, the Falklands, New Zealand, and southeastern Australia. The common diving petrel (*P. urinatrix*) is circumpolar; the largest species, the Peruvian (*P. garnotii*), has followed the Humboldt current along the west coast of South America and breeds from Chile to Peru; the Magellanic (*P. magellani*) is confined to the tip of South America; the Georgian (*P. georgicus*) breeds on South Georgia, Macquarie, and Auckland islands. Diving petrels are specialized birds with short, black-and-white bodies and closely resemble externally and in habits the small auks (family Alcidae, order Charadriiformes) of the Northern Hemisphere.

NATURAL HISTORY

Locomotion. All tube-nosed birds have a protracted life cycle conditioned by their evolution and oceanic environment. Because they spend most of their lives at sea, they are clumsy on land, laboriously using their wings as props to assist locomotion; their legs are too far to the rear to effect a well-balanced bipedal progress. The smaller species nest in burrows and rock crevices and are nocturnal, being helpless and unable to manoeuvre quickly on land when attacked by predators. As a rule, the incubating bird is tame and does not associate the approach of man with danger but will often allow him to stroke and fondle it. Albatrosses are especially docile, hence the name mollymawk (mallemuck), from the Dutch *mollemok* ("stupid gull").

Flight pattern of albatrosses

The long-winged tubinares require a smooth runway for takeoff on a calm day; over rough ground they will utilize the bill to hook along, and either climb a rock or tree to gain a launching height or flop over the edge of the nearest cliff. On the wing, they are perfect aviators, riding out the severest storms of their normally windy oceanic feeding grounds with ease and grace. The great albatrosses can overtake and circle a fast ship at sea, with long glides rarely interrupted by wingbeats. The ability of albatrosses to move upwind without flapping depends on the fact that wind velocity is appreciably lower near the waves than a few metres in the air. The flight pattern is a series of broad ellipses (Figure 12), characteristically including (1) a fast downwind glide beginning at the greater elevation, at which the wind velocity is higher, giving the bird considerable momentum with a small loss of altitude. As it reaches maximum speed, the bird turns crosswind (2), briefly skimming the waves, then moves upwind (3) on a long, level glide through the lower velocity air, maintaining altitude, but losing speed. As soon as its air speed reaches a critical low, the bird climbs steeply crosswind (4) to complete the ellipse, now at the same altitude as at the start of the previous downwind leg but having gained in position upwind. It then glides downwind again to gather fresh momentum. The same flight pattern may be used, of course, to travel crosswind or downwind. The normal air speed of the royal and wandering albatrosses (*Diomedea epomophora* and *D. exulans*), whose wing spans reach about 3.4 metres (11 feet), is 80 to 110 kilometres (50 to 70 miles) per hour. Although the flight appears effortless, some energy is expended in the muscular action that keeps the long, narrow wings fully extended.

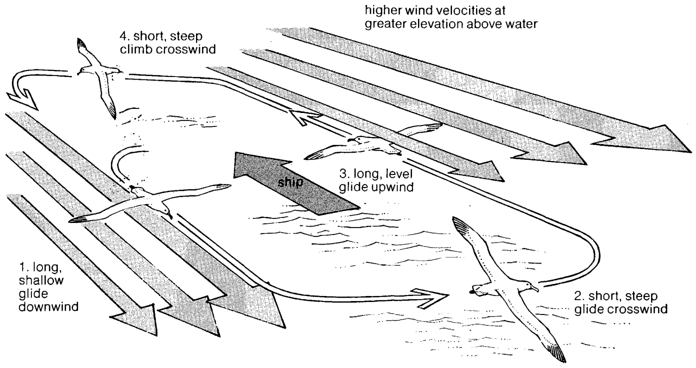

Drawing by R. Keane

Figure 12: Flight pattern of the albatross.

The medium-sized tubinares (shearwaters and procellariid petrels) have a flight pattern similar to that of albatrosses, but their shorter wings are flapped regularly between the briefer gliding periods. The little storm petrels have an altogether more erratic darting, fluttering, and sometimes hovering flight, their feet hanging down to walk on the water surface.

Feeding habits. Shearwaters, storm petrels, and diving petrels feed by taking small fish and crustaceans close to the surface; they make short dives as necessary. Many of the larger procellariids consume substantial amounts of squid. Albatrosses, giant petrels, and fulmars dive little; they are surface feeders, often settling on the water. At night they devour squid that rise to the surface; during the day they take schooling fish; garbage from ships; wounded, exhausted, or dead birds; and carrion, including the flesh of dead whales and other cetaceans. The giant petrel is probably the only tubinare agile enough on land to kill other birds; at its nesting grounds it will attack young penguins inadequately guarded by their parents.

Reproduction and growth. As a general rule, the mature adults return to the established breeding site many weeks before the single white egg is laid. There is often severe competition for nesting territories in large crowded colonies on small islands. Returning each year to the same nest site, the male and female remain faithful to it and thus to each other for life. It is believed that some albatross pairs also remain together at sea in the nonbreeding season, but many of the burrowing shearwaters and petrels, meeting on land only at night, may never see their mates clearly when ashore (recognition being by voice, touch, and possibly smell) and probably do not deliberately consort in pairs at sea.

The greeting ceremony

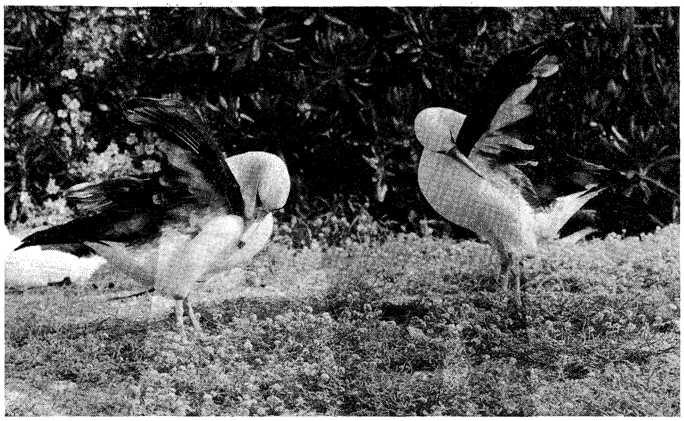
At each fresh encounter ashore between breeding birds, there is an elaborate greeting ceremony; the birds clash and fence with the bills, cackling and screaming. These antics occur in both the nocturnal procellariids and the diurnal albatrosses, and in the latter there is also a bowing and dancing display. Such behaviour provides time for mate recognition and relieves and displaces any natural aggression or fear.

The nest type varies somewhat among species. Albatrosses scrape a shallow depression or build a mound of soil and vegetation; the fulmars and other diurnal procellariid petrels nest on ledges or on level ground; most shearwaters, diving petrels, and some storm petrels dig burrows in soft soil; other storm petrels utilize natural crevices.

Once the nest site has been adopted, one member of the pair usually remains on guard against usurpation by other birds that are still in search of a suitable nest site. The male may remain on guard for several days and nights, while the female feeds at sea in order to meet the nutritional requirements of the developing egg. In some species, the female may depart on a recuperative feeding cruise within a few hours after laying if her mate is there to take over incubation. Mated birds do not feed each other; instead they incubate for periods of several days each, the bird at home fasting and losing weight while the bird at sea is feasting and fattening.

The egg is incubated for a long period, about 80 days in the wandering albatross, 52 days in the Manx shearwater, 40 days in the smallest petrels (the last about equal to the incubation period of the ostrich). For the first week or so after hatching, the helpless, downy chick requires the warmth of the parental body for survival. During this period it is brooded and fed tenderly on an oily broth of semidigested marine organisms pumped from the adult esophagus, which is muscularly constricted to control the flow to the infant's needs. Instinctively, the chick seeks the open, warm, fishy-smelling mouth of the parent, thrusting and groping blindly with its tiny bill crosswise in the open maw of the adult.

Olin S. Pettingill, Jr.—The National Audubon Society Collection/Photo Researchers



A pair of Laysan albatrosses (*Diomeda immutabilis*) in ritualized preening display, a part of the courtship that precedes nest building.

The down grows rapidly; a second down sprouts (to which the first remains attached), and the baby is soon homeothermic—*i.e.,* able to keep itself warm while its parents forage far at sea and return with increasingly large crop loads of food. Some of the fishing trips of the parents are real voyages, involving absences of several days; the Manx shearwater may travel nearly 1,000 kilometres (more than 600 miles) from Wales to the Bay of Biscay and back to load up with its favourite food, sardines. Albatrosses may leave their well-developed nestling for a week or two. If both parents happen to return at the same time, the nestling may ingest food equal to its own weight in one meal. It becomes very fat in the later stages of the long fledgling period, which is not less than two months in the small petrels and reaches nine months in the largest albatrosses. Before it leaves the nest, the chick is deserted by the parents, who retire to molt at sea. This begins a starvation period, which may last a week in the smallest petrels, 12 days in the medium-sized shearwaters, and considerably longer in the largest species, before the fledgling goes to sea. When deserted, it is well-feathered and fatter and heavier than the adult; it needs a period of thinning and exercise before it is capable of flight. After days of fasting and wing flapping, it may become airborne one windy night, especially if hatched in a burrow on a gale-swept mountain height from which it can flap and glide to the sea. Calm weather is its enemy; many island-born young tumble down to the sea, too heavy to take off again in still air. They are expert swimmers, however, and can dive deeply to avoid attacks by aerial predators.

Period of fledgling starvation

Paddling rapidly away from the dangers of land and soon gaining flight, the young procellariiform sets off along the traditional migration route, alone and unguided by the long-departed adults. Driven by an innate impulse to keep flying, it reaches winter quarters that it has never seen before, often at a surprising speed. One Manx shearwater, banded in Wales as a fledgling, travelled 9,900 kilometres (6,200 miles) to southern Brazil in 16½ days. Allowing half of each day for resting and feeding, this is equivalent to an average surface speed of 50 kilometres (30 miles) per hour over the period, a remarkable achievement for a bird just out of the nest.

The young albatross remains longest in the nest—so long in the case of the royal and wandering albatrosses that the nestling is overtaken by the Antarctic winter. It endures blizzards and savage winds that force it to grip the nesting mound tightly with its claws, yet it is warm enough under its oily plumage to survive fasting for many wintry days until a parent appears with food. Because of the protracted nesting period, these great albatrosses cannot rear more than one young every other year. To compensate for the slow rate of reproduction, albatrosses are long-lived; life expectancy, once breeding age is reached, appears to be several decades. Marking of Laysan albatrosses has shown that they do not breed successfully until seven years old. In order to maintain their numbers, the wandering and royal albatrosses, breeding for the first time even later, must have the highest average longevity among birds. The Manx and other medium-sized shearwaters lay the first egg when five years old, the least petrel in the third or fourth summer.

There is thus always a large proportion of each tubinare population that is nonbreeding. During its first year at sea, the young bird may not even approach the land. While the mature birds have completed their migration and settled to breed at home, the yearlings may lag far behind on the route, spending the summer at sea. In the next few years adolescents arrive at the breeding islands and shores too late to make more than a preliminary landing and exploration of the ground for a future partnership. At midsummer in the nesting colonies there is a considerable arrival of immature birds familiarizing themselves with prospective breeding territories. Where a colony is already overcrowded, it is always the young, eligible, but inexperienced, birds that leave to form new peripheral colonies in the region of their birthplace.

Ecology and conservation. The tubinares collectively occupy a position midway in the oceanic food chain. Their food is lower on the chain than that of animals that take subsurface fishes (*e.g.*, certain diving birds, seals, and medium-sized predatory fishes) but above that of the minute and larval fishes which feed on plankton. Insofar as man has overfished much of the oceans and removed or reduced competitors, such as the krill-feeding whales, and has taken fish species from the food chain, he might be said to have influenced the success and therefore the numbers of the tubinares, but there is no firm evidence for this. Although the numbers of some species of tubinares are recorded as increasing, it is likely that the reason is stricter protection and reduction of exploitation at the nesting sites. It is generally accepted that the considerable increase of the fulmar in the North Atlantic and its spread south from the Arctic to new breeding grounds has resulted from the great development of fishing fleets, which fulmars follow in vast numbers to devour the fish offal thrown overboard.

Man's effect on the tubinare population

Although several tubinares are considered rare and a few in danger of extinction, the majority are successful and often numerous. With their pelagic, aerial habits they are less liable to be trapped in oil slicks than are the Northern Hemisphere swimming seabirds (auks, ducks, and loons), which are killed by the thousands by oil. The diving petrels and prions, living in the cleaner Antarctic seas, are less exposed to this danger.

FORM AND FUNCTION

General features. The general body plan of procellariiform birds varies slightly from family to family. In general, they are long-winged, short-necked birds with short to moderate tails and legs. Webbing is present between the front toes, and the hindtoe (hallux) is small or lacking. In contrast to their strong-flying relatives, the diving petrels have short wings. At the other extreme, the aspect ratio (the ratio of wing span to the chord, or width) of the wing may exceed 14:1 in some albatrosses. This long, narrow wing, with a high-lift airfoil, is an extreme adaptation for fixed-wing gliding.

The bill varies from rather short and broad in diving petrels to medium in length (somewhat more than half the total length of the head) in some albatrosses. It is sheathed in horny plates and has a distinct hooked nail at the tip. In albatrosses the two nasal tubes lie separated on the right and left upper lateral surfaces of the bill; in all other procellariiforms, the nostrils are fused into a single tube lying on the dorsal midline of the bill and having a dividing wall or septum, which may end short of the end of the tube, resulting in a single opening.

Drawing by R. Keane

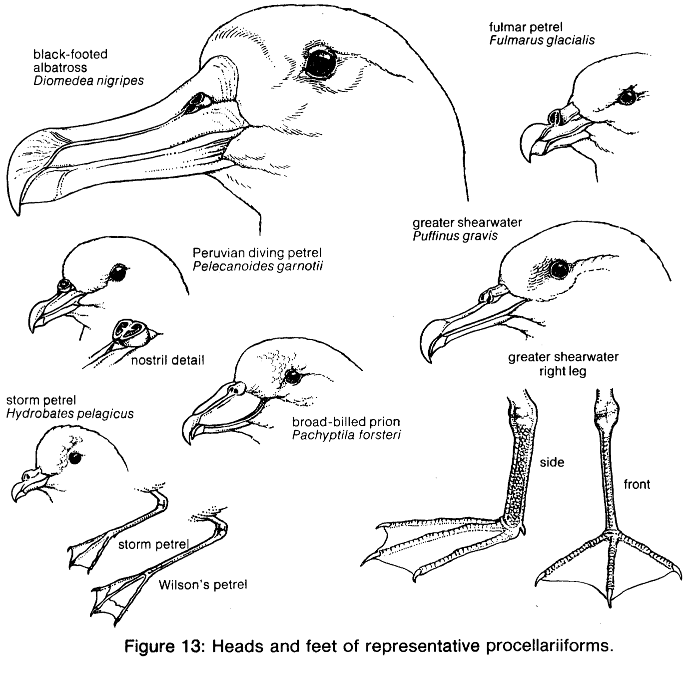


Figure 13: Heads and feet of representative procellariiforms.

Procellariiforms are totally lacking in bright plumage colours, being entirely black, white, or shades of brown or gray. Strikingly contrasting patterns of light and dark are often found, however, and the bills or feet of a few species are yellow or pink. A number of shearwaters and procellariid petrels and a few albatrosses are polymorphic; *i.e.*, they occur in light and dark phases (plumage types), some species also having intermediate forms. The polymorphism may be restricted to certain parts of the plumage, such as the underparts of the body or the upper surface of the wings.

Squirting oil as a defense mechanism

Stomach oil. Most tubinares, when handled or threatened, eject the oily contents of the stomach with some force. In some species, notably the cliff-nesting fulmars, this habit, a fear reaction that also serves to lighten the bird for flight, has been exploited as a defensive weapon. Facing an intruder, the disturbed bird ejects a spurt of evil-smelling fluid a metre or so in his direction, often with apparently planned accuracy. The habit is instinctive; a baby fulmar, on hatching, has been observed to squirt yellow oil before it is fully out of the shell. Later, the downy chick squirts oil at any visitor, even its parents. Mated fulmars may exchange little squirts of oil during the excitement of bill-fencing ceremonies.

Analysis of this unique oil shows that it is a waxy secretion of the proventriculus (the first chamber of the stomach), rich in vitamins A and D. In most birds, the walls of the proventriculus produce an acid fluid that rapidly breaks down raw food entering from the esophagus. In the tubinares, which feed their young a soup of predigested marine organisms, the proventriculus is much enlarged and internally folded, increasing the surface when dilated and enabling a larger number of glands to function. The latter are groups, or follicles, of oil-producing cells. The colour of the oil varies according to the type of food; it is often reddish from the presence of astacin, a pigment found in crustaceans.

The discharge of stomach oil is partly excretion of surplus fat, which might upset the bird's metabolism if retained in quantity. Ejected through the mouth and nose, it also disposes of excess vitamins and salt in the diet of marine food and seawater. Similar in character to the secretions of the oil glands of other birds, the crop oil may also assist in waterproofing the feathers as the tubinare preens its plumage with its oil-stained bill.

EVOLUTION AND CLASSIFICATION

Relationship to penguins

Evolution. The oldest tubinare fossil is a giant albatross (*Gigantornis*) from the Eocene of Nigeria (about 50,000,000 years ago). It may have had a wing span of six metres (20 feet) and was contemporary with the now extinct giant penguins (order Sphenisciformes). It is generally agreed that the two orders had a common ancestor from which they may have evolved. The penguins occupied the ecological niche of diving and feeding under the surface and became flightless; the tube noses specialized in flight and surface feeding. Support for a common origin (a view contested by some taxonomists) comes from the facts that the oldest fossil penguin had a bill with distinct tube-nosed apertures; the young of the blue penguin *Eudyptula*, considered to be the most primitive of penguins living today, exhibits tubelike openings to its nostrils. Mutual displays of bill fencing and wing movements in courtship, as well as the method of regurgitating digested food, are almost identical in tubinare and penguin. The short-winged diving petrels, which "fly" much underwater but little in air, seem to parallel an early stage in the evolution of penguins, especially when, during a few weeks of the annual molt, they lose their quills and must live in the water.

Classification. *Distinguishing taxonomic features.* The families of the Procellariiformes are separated mainly by the general body plan, the condition of the nostrils, and, in the case of the Procellariidae and Hydrobatidae (long considered one family), the osteology of the skull and sternum. At the genus level, characters used include the shape of the beak, wings, and tail; the degree of flattening of the tarsus (lower leg); size of the hallux; and the relative lengths of the leg bones.

Annotated classification. The following classification, proposed by American ornithologist Alexander Wetmore in 1930, is in nearly universal usage.

ORDER PROCELLARIIFORMES (tubinares)
Oceanic birds with tubular nostrils; bill covered with horny plates and hooked at the tip. Anterior toes webbed; hallux short or lacking. Wing with 11 primary feathers (the outer minute); secondaries short; diastataxic (*i.e.,* with the 5th secondary absent). Two coats of nestling down. Oil gland feathered. Strong musky smell. Single white egg; long incubation and nestling periods. About 87 living and 36 known fossil species; all marine; worldwide.

Family Diomedeidae (albatrosses)
Middle Eocene to present. Extremely long, narrow wings; short tail. Bill longer than remainder of head; nostrils semitubular, small, situated near the base of long groove. Two genera, 13 species; length 50–125 cm (20–50 in.); wing span to 340 cm (11 ft); North Pacific and all southern oceans.

Family Procellariidae (large petrels, fulmars, prions, shearwaters)
Middle Oligocene to present. Long-winged, short-tailed. Nostrils united on top of bill. Twelve genera, about 50 species; length 22–75 cm (9–30 in.); all oceans, but greatest diversity in Southern Hemisphere.

Family Hydrobatidae (storm petrels)
Upper Miocene to present. Small black and brown birds, usually with conspicuous white rump; wings rounded; tail square or forked. Often walk on water. Eight genera, 20 species; length 15–20 cm (6–8 in.); all oceans, but more species breeding in Southern Hemisphere.

Family Pelecanoididae (diving petrels)
Upper Pleistocene to present. Small stocky birds, with short wings and tails. Black above; white below. One genus, four species; length 16–20 cm (6½–8 in.); cool subantarctic seas.

Critical appraisal. In a system positing evolutionary relationships on the basis of structural affinities, the tube-nosed birds seem to fit conveniently between the penguins and the pelecaniform birds. But within the order, the arrangement of genera, species, and races of tubinares has been changed all too frequently by taxonomists. This disorder has resulted partly from new information on the structure, habits, and distribution of these birds, hitherto little studied because of their remote oceanic breeding grounds. The ordinal position occupied by the highly specialized prions and diving petrels is still in dispute. The fulmars are sometimes divided into three species because of their distinct differences in bill, body size and shape, and geographical range; but other taxonomists regard them as Arctic, Pacific, and Antarctic races of the species *Fulmarus glacialis.* Several authorities have recently placed several medium-sized shearwaters in one species *Puffinus puffinus,* with eight closely related forms or subspecies. Because there is no evidence of interbreeding between these populations, although they may mingle at sea on migration, and there are no intermediate forms even when they inhabit the same island (in which case, they nest at different times), other workers prefer to treat them as full species of *Puffinus.*

(R.M.L.)

Sphenisciformes (penguins)

The penguins are a group of flightless seabirds constituting a particularly distinct and homogeneous order, Sphenisciformes; the one family, Spheniscidae, is limited in distribution to the Southern Hemisphere. Although a few penguins inhabit temperate regions and one, the Gálapagos penguin (*Spheniscus mendiculus*), lives at the Equator, the majority of the 18 species breed on islands in subantarctic waters.

GENERAL FEATURES

The stocky, short-legged appearance of penguins has endeared them to the popular mind. They range from about 35 centimetres (14 inches) in height and approximately one kilogram (about two pounds) in weight, in the little blue, or fairy, penguin (*Eudyptula minor*), to 115 centimetres (45 inches) and 25 to 40 kilograms (55 to 90 pounds) in the emperor penguin (*Aptenodytes forsteri*). Most are black on the back and white below, often with lines of black across the upper breast or spots of white on the head. Colour is rare, being limited to red or yellow irises of the eye in some species; red beaks or feet in a few; yellow brow tufts in the three species of *Eudyptes;* and orange and yellow on the head, neck, and breast in the two species of *Aptenodytes.*

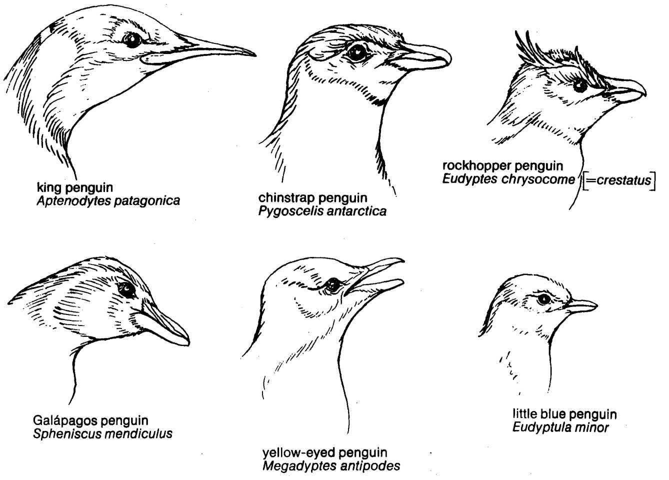


Figure 14: Heads of representative sphenisciformes.
Drawing by Christian D. Olsen

Breeding colonies

The total populations of some species, such as the emperor penguin, are estimated in hundreds of thousands, but those of most species of smaller penguins certainly run to several million. The largest breeding colonies are on the islands between 50° south latitude and Antarctica. These immense colonies, some of which contain hundreds of thousands of individuals, represent a large potential food resource, but the economic importance of penguins is negligible. Nineteenth-century whalers and seal hunters visited some colonies for meat and eggs, and a penguin oil industry once took large numbers of birds; by the early 20th century, however, this exploitation was no longer profitable, and most colonies were left alone or actively protected. Some species are now increasing in numbers, apparently as a result of the drastic reduction of Antarctic whales, which compete with penguins for the krill (minute crustaceans) on which both feed.

NATURAL HISTORY

Reproduction. Many features of the life cycle vary with the body size of the species and its geographical distribution; the chronology of breeding may also vary within a species in relation to latitude. Certain species, such as the Cape or jackass penguin (*Spheniscus demersus*), probably other members of this genus, and the little penguin, breed twice a year, but the majority of species breed only once each year and the king penguin (*Aptenodytes patagonica*) twice in three years. One egg is laid by the emperor and king penguins, and two, or occasionally three, by all others. Most penguins begin breeding in the austral spring or summer; king penguins are on a 14- to 18-month cycle, and the timing of an individual pair depends on the success or failure of the previous breeding attempt. Some populations of the gentoo penguin (*Pygoscelis papua*) also breed in winter. The breeding of the emperor penguin begins in autumn, apparently timed so that the long developmental period will produce the young in midsummer, when their chances of survival are greatest.

The gentoo, which has a circumpolar distribution, is notable for the lack of synchrony among populations, but otherwise its breeding schedule is essentially comparable to that of most penguins. In the Crozet Islands, off southern Africa, for example, egg laying takes place in July. The two eggs are incubated for 35 or 36 days, and the rearing of the chick takes two months. The last immature birds go to sea in January, leaving the adults to molt in February.

Many types of visual and vocal displays are employed between the arrival of the birds at the colony and their departure. Courtship calls are used during pairing and to a lesser degree during the succeeding phases of breeding. There is marked vocal dimorphism between sexes in the emperor penguin and the king penguin, and less marked dimorphism in some other species. Upon arrival at the colony each bird returns to the nest that it left the previous year and generally rejoins its mate of the previous year, unless the death of the latter forces it to choose another partner. This applies even to the emperor penguin, which is capable of finding its mate despite the absence of a nest and the large size of the colony.

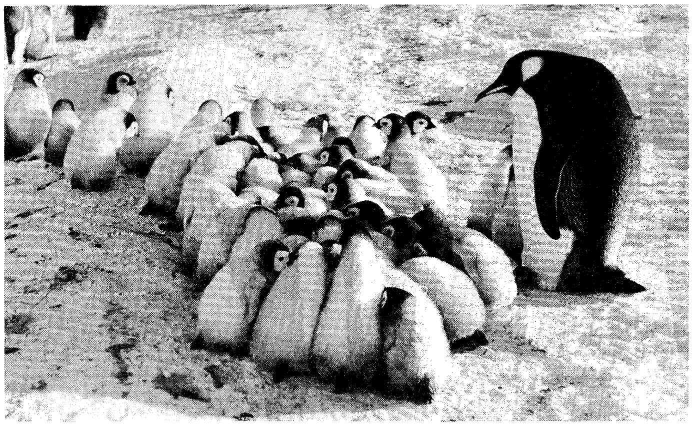
Types of display

The displays that occur with the reassembly of the colony and the finding of mates, as well as those preceding copulation, are quite similar among the majority of species, but the accompanying vocalizations are more diverse. Various species have been described as braying, trumpeting, croaking, cackling, and cooing. The behaviour of experienced older birds is more elaborate and more effective than that of younger individuals. Adélie penguins (*Pygoscelis adeliae*) may return to the reproductive colony from their third year onward but do not breed successfully until their fifth or sixth year.

Incubation of eggs is performed by both sexes in all species except the emperor penguin, in which it is done by the male alone. The incubation period begins immediately following the laying of the single egg in the emperor and king penguins and after the laying of the second egg in other penguins. With the advent of incubation, the bustle and myriad cries that characterized mating give way to quiet and inactivity. Faulty incubation behaviour by inexperienced birds frequently results in the abandonment or breakage of eggs. The mortality rate (eggs and chicks) is very important at the egg stage, varying from year to year depending on climatic conditions, the percentage of young birds in the reproductive population, and the pressure of predation. In general, the mortality (eggs and chicks) is from 40 to 80 percent of the eggs laid. In coastal colonies predators include, in order of importance: skuas (*Catharacta*); sheathbills (*Chionis*); and the giant petrel (*Macronectes giganteus*). On the Australian, African, and South American continents, the nocturnal habits of certain penguins and the fact that they nest in burrows substantially limit predation, which is mostly by gulls and man.

Following egg laying, the female usually departs for the sea to feed, returning to relieve her mate after about 10 to 20 days, depending on the species; thereafter, they alternate in periods of a week or two. The female emperor penguin, however, must often walk 80 to 160 kilometres (50 to 100 miles) from colony to the sea and does not return until the end of the incubation period. During the 64-day incubation period, which extends through the height of the Antarctic winter, the male emperor penguin incubates the egg, holding it on his feet and living on stored fat reserves. During violent winter storms, members of the colony gather for mutual protection from wind and cold in tightly packed crowds called huddles.

Jean Prevost



Crèche of emperor penguins (Aptenodytes forsteri).
To guard against predators, the young are herded into groups attended by adults while the parents fish for food.

Emergence from the shell takes 24 to 48 hours, during which the brooding parent is particularly irritable. The chick shows feeding behaviour immediately on hatching, receiving a regurgitated "soup" of crustaceans or fish by inserting its bill into the open mouth of the parent. During its early days the young bird is sheltered under the body of one of its parents, who take turns foraging and brooding. Growing larger, the young bird remains at a parent's side, although able to maintain its body heat and move about alone. In surface-nesting species, following the individual rearing phase in the nest or (in the emperor and king penguins) on the feet of the parent, the chick joins 100 or more of its contemporaries in a nursery group, or crèche, sometimes guarded by a few adults, while both its parents forage at sea. Upon returning with food, the parent calls its chick from the crèche and is able to distinguish it from other chicks (which frequently also respond) by voice and appearance. During the breeding season the number of "unemployed" adults in the colony increases with the addition of those who have lost eggs or chicks. In emperor penguin colonies, these unemployed birds often interfere with parents that have young and cause increased mortality. During the crèche stage the fuzzy down that has covered the chick since hatching is replaced by a coat of short stiff feathers, like those of the adult but usually somewhat different in colour. Once this molt is complete, the juvenile leaves the colony to seek its own food at sea.

Crèche behaviour

The period of growth of the young bird from hatching to complete independence varies from two months, in the smallest species of the genus *Eudyptula,* to $5^1/_2$ months in the emperor and 12 to 14 months in the king penguin.

Molt. Adult penguins molt all of their feathers once a year, following the breeding period. While in molt the bird is unable to enter the water and instead withdraws to a communal molting site usually situated in a sheltered area away from the colony. The duration of the molt varies from about two weeks in small species to more than a month in the larger ones.

Locomotion and orientation. Penguins are adapted for rapid locomotion in water, in which the wings, or flippers, are used for propulsion; the birds "fly" under water. When moving at high speed they frequently leave the water in leaps that may carry them a metre or more through the air; during this time they breathe.

On land, penguins are much more awkward, and their rocking walk is highly amusing to watch. Despite their short legs, however, penguins can run with surprising speed. Some species, such as the rockhopper (*Eudyptes chrysocome*) and Adélie penguins, move among rocks with agility, using the flippers for balance. When moving rapidly on snow or ice, many penguins "toboggan," sliding on the belly as they propel themselves by the feet and flippers. The flippers, along with the beak, are the prime weapons in defense and attack.

"Tobogganing"

Scientists have long wondered how penguins are able to find their way back to their colonies from far out at sea, where currents may have carried them great distances, and how they are able to direct themselves correctly on land, in the absence of clear-cut landmarks. Recent studies have shown that penguins transported into the interior of Antarctica are able to find their way back to the ocean by using the sun as a directional aid. It is probable that the same means of orientation is used at sea, and that upon approaching the coast they are able to recognize features of the shoreline and the ocean bottom.

Food habits. The type of food utilized varies with the species, the geographical region, and the time of year. Most of the smaller southern penguins feed primarily upon krill (crustaceans of the family Euphausidae), which attain high densities in the rich, well oxygenated Antarctic waters. Cephalopods (squid and cuttlefish) and small fishes may form substantial fractions of the food and, in a few, such as the Cape penguin, comprise the basic elements of the diet. The total weight (biomass) of food consumed by a large penguin colony is prodigious, often exceeding several tons per day.

Natural enemies. The principal enemies of penguins at sea are the leopard seal, or sea leopard (*Hydrurga leptonyx*), and the killer whale or grampus (*Orcinus orca*). The eared seals (*Otariidae*) of the genera *Arctocephalus* and *Neophoca* also take penguins in the Australian–New Zealand and subantarctic regions.

FORM AND FUNCTION

The penguins are highly specialized for their flightless aquatic existence. The feet are located much farther back than those of other birds, with the result that the bird carries itself mostly upright; its walk can thus be described as plantigrade (*i.e.,* on the soles). It rests on the ground on a sole, comprising the whole foot, instead of one formed of the toes, as in other birds. The most notable character of the group is the transformation of the forelimb into a swimming member, accompanied by a body morphology particularly adapted to movement in a liquid medium. The thoracic (rib) cage is well developed, and the sternum bears a pronounced keel for the attachment of the pectoral muscles, which move the flippers. The flipper has the same skeletal base as the wing of flying birds but with its elements shortened and flattened, producing a relatively rigid paddle-like limb covered with very short feathers, which makes an ideal organ for rapid propulsion. The body plumage likewise consists of very short feathers, which minimize friction and turbulence; by their density and the layer of air which they retain, the feathers provide almost ideal insulation of the body.

Structure of the forelimb, or flipper

The insulation of the bird's body is particularly important for Antarctic species that move in water that is always below 0° C (32° F). The cooling power of seawater at −1.9° C (28.6° F) is equal to that of a temperature of −20° C (−4° F) with a wind of 110 kilometres (70 miles) per hour. The velocity of movement of the birds in the water serves to increase this heat loss. In fact, the skin is insulated by a layer of air trapped under the plumage, and the only bare skin in direct contact with the water is that of the feet. In the case of the emperor penguin on land in winter, the feet may also be in close and permanent contact with ice. The skin temperature is then in the neighbourhood of 0° C, and snow does not melt upon contact. This low temperature is made possible by remarkable anatomical arrangements in the lower limb, in which closely adjacent arteries and veins form a system of heat exchange between opposing flows; in the upper member a remarkable heat-exchange network permits the cooled returning blood to absorb heat from the outflowing blood, providing maximum economy of heat consistent with the functioning and survival of the foot.

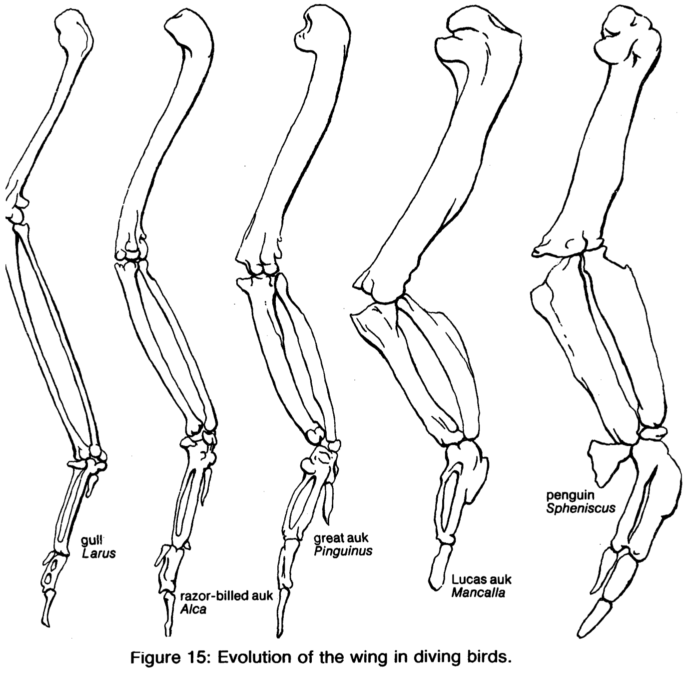

From R.W. Storer, "Diving Birds," International Ornithological Congress (1958)

Figure 15: Evolution of the wing in diving birds.

Like the petrels, the penguins have supraorbital glands or salt glands, which enable them to ingest hyperosmotic substances (*i.e.,* those containing more salt than the body tissues); excess chloride is excreted in the form of a solution the concentration of which is greater than that of seawater. These glands are already functional in the young penguin chick, which begins to consume food of marine origin from its first day of life.

Recent research has shown that the species most isolated geographically, such as the emperor penguin, can be subject to viral diseases of the psittacosis–*Lymphogranuloma* type. Some, such as the Adélie penguin, carry in their bodies trace amounts of pesticides such as DDT, albeit in lower quantities than are found in many birds that live closer to man.

EVOLUTION AND CLASSIFICATION

Fossil record. Evidence from paleontology indicates that the penguins and tubinares (Procellariiformes: albatrosses, shearwaters, and petrels) had a common origin. Both groups are represented by well-defined fossils from the Eocene (about 50,000,000 years ago). During the Tertiary Period (from about 65,000,000 to 2,500,000 years ago) the sphenisciform line produced a number of distinctive side branches, all recognizably penguins. From the taxonomic point of view the radiation (evolutionary differentiation) had reached the level of the subfamily in the Miocene (from 26,000,000 to 7,000,000 years ago), and the diverse subfamilies of that epoch have been replaced in the present-day group by one in which the radiation has reached the level of genera. All of the fossil remains of penguins have been collected within the zone of the present-day distribution of the Sphenisciformes. Some apparently lived in warmer regions than do most of today's penguins. Certain species, such as *Palaeospheniscus patagonicus,* appear to have had a height intermediate between the large and the medium-sized modern species; others, such as *Pachydypees, Anthropornis,* and *Palaeeudyptes,* reached heights of between 120 to 160 centimetres (approximately 50 to 60 inches).

Classification. One authority has divided the *Spheniscidae,* fossil and modern, into five subfamilies: *Palaeospheniscinae* and *Paraptenodytinae* of the Patagonian Miocene; *Anthropornithinae* of New Zealand, Patagonia, and Antarctica; *Palaeeudyptinae* of New Zealand and Australia; and finally *Spheniscinae,* which includes all the present-day genera. (J.Pr.)

Pelecaniformes (pelicans, boobies, tropic birds, cormorants, frigate birds)

The order Pelecaniformes comprises a number of diverse groups of aquatic birds that share the common characteristic of webbing between all four toes. There are three main groups (suborders): Phaethontes, the tropic birds; Fregatae, the frigate birds; and Pelecani, the pelicans, boobies (including the gannet), cormorants, and anhingas (or darters). All are relatively large birds: they range in length from about 40 centimetres or 16 inches (excluding the elongated central tail feathers) in the white-tailed tropic bird (*Phaethon lepturus*) to 1.8 metres (six feet) in the Dalmatian pelican (*Pelecanus crispus*).

GENERAL FEATURES

In terms of their way of life, the pelecaniform birds fall into four adaptive groups: frigate birds—long-winged masters of piracy and aerial pursuit of surface-living marine prey; tropic birds and boobies—wide-ranging flyers that capture prey underwater by plunging from a height; pelicans—large, large-billed, long-necked, buoyant birds that fish mostly by reaching down while swimming at the surface; cormorants and anhingas—heavy-bodied, long-necked, underwater swimmers, respectively pursuing and lying in wait for their prey below the surface. The anhingas are largely confined to freshwater; pelicans and cormorants occur in both freshwater and marine habitats; and the other groups are entirely marine.

Production of guano as fertilizer

Importance to man. Most pelecaniform birds are of rather little significance to man, but the guano (excrement) of cormorants, boobies, and pelicans is an important fertilizer. Exploitation of old accumulations of guano reached its peak in the mid-19th century, and since then only the current production of guano has been available in most areas, but even this provides a substantial resource where the bird populations are large. There were an estimated 18,000,000 guano birds on the coast of Peru early in the 1960s. Of these, about 15,000,000 were guanay cormorants (*Phalacrocorax bougainvillii*), and the remainder were Peruvian boobies (*Sula variegata*) and brown pelicans (*Pelecanus occidentalis*). The harvest of guano at that time amounted to about 180,000 tons per year. The guano birds in the area feed largely on the Peruvian anchovy, and now that this fish is directly exploited on a large scale for fish meal and fish oil, there is some competition between the birds and the fishing industry. In Southwest Africa artificial platforms have been constructed in coastal lagoons and on an offshore reef, greatly facilitating the collection of guano.

Effects of insecticides

Pelicans, feeding on fish from inland and coastal waters, are among the animals whose diet tends to ensure that they will accumulate residues of insecticides (especially DDT) in their bodies. It recently has been found that among the physiological effects of these substances on birds are changes in calcium metabolism that result in their laying eggs with abnormally thin shells, or no shells at all, with the result that these eggs usually break before hatching. Effects of this kind have entirely prevented the successful reproduction of brown pelicans on the California coast during recent years, and some populations elsewhere are also showing adverse effects.

Drawing by R. Keane

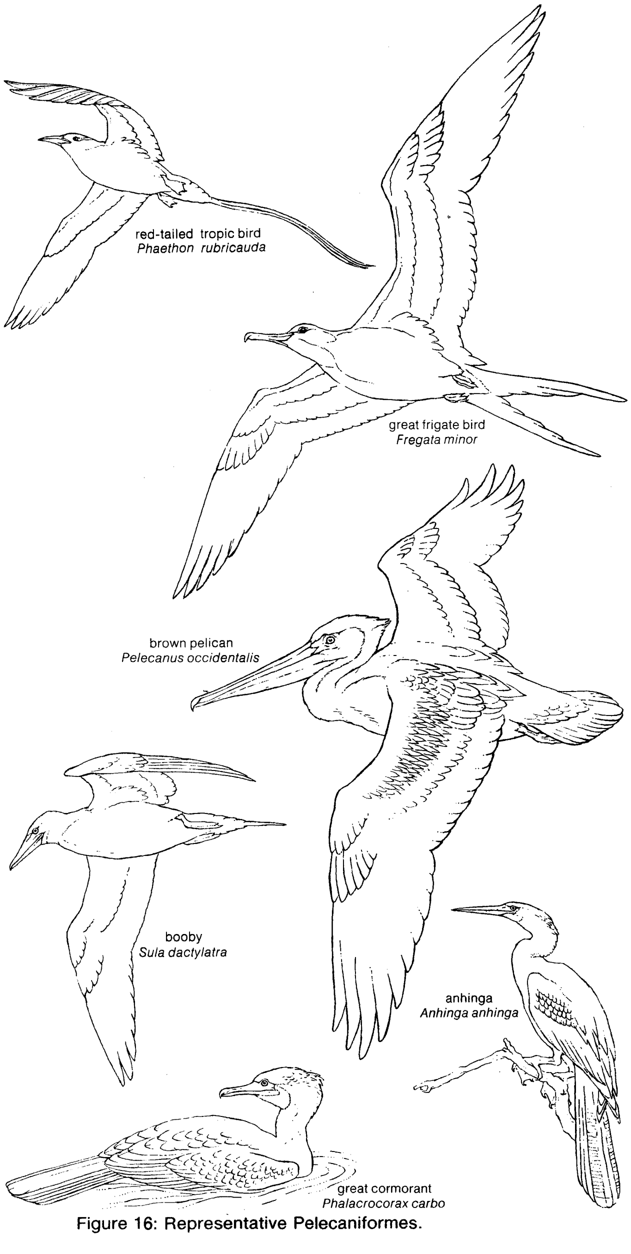


Figure 16: Representative Pelecaniformes.

Distribution and abundance. Pelecaniform birds occur all over the world except in parts of the interior of North America, Africa, Asia, and Australia, in the high Arctic, and in most of Antarctica. Two groups, the tropic birds and frigate birds, are essentially confined to the tropics; the boobies, pelicans, and anhingas are widespread in the tropics but also penetrate far into the temperate zones; and cormorants breed from the Equator to the Arctic Circle and the Antarctic Peninsula.

The tropic birds are the most pelagic (free-flying over open ocean) of the pelecaniforms and can be seen even in the most unproductive central parts of tropical seas. Some boobies and frigate birds are found far out at sea, but most of them return to land to roost at night. Pelicans, cormorants, and anhingas do not venture far from land but may commute some distance between roosting or breeding places and their feeding grounds.

NATURAL HISTORY

Breeding grounds

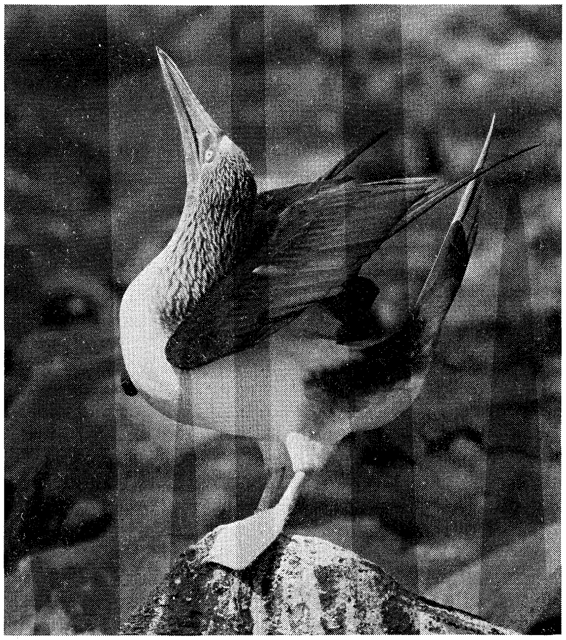
Reproduction. The breeding of pelecaniform birds is essentially restricted to places free of mammalian predators. Tropic birds, boobies, and frigate birds typically breed on oceanic islands or on small islets and stacks off continental coasts. When man has introduced predators (such as cats) to isolated islands, pelecaniforms and other marine birds have often been eliminated from their traditional breeding grounds. On Ascension Island and St. Helena in the South Atlantic, for example, tropic birds, frigate birds, and boobies, which used to breed in large numbers on the main islands, have either been exterminated or have been forced to confine their nesting to sites on small offshore islets. Pelicans and cormorants are not normally found on islands far from continental land but breed on islands in lakes or offshore, or in other protected sites such as trees standing in water or on cliffs. Anhingas breed in trees or bushes close to the sheltered waters where they prefer to feed.

Pelecaniform breeding, whether on the ground or in trees, is typically colonial, apparently because of the scarcity of safe places. If they are to breed at all, all the birds of an area must crowd into the available space. In practice, most colonies have a fairly definite upper limit of density determined by the distance that an incubating bird can reach out to repel intruders. Nests of the gannet (*Morus bassanus*) have an average density of about one per square metre; those of the guanay cormorant average about 3.5 per square metre.

In some species, colonial breeding has become obligatory, and single pairs or small groups do not breed successfully. Other species breed colonially only where there is a shortage of space for nesting. The masked booby (*Sula dactylatra*), for example, breeds in dense colonies on islets off Ascension Island but in dispersed patterns on Christmas Island (Pacific). Breeding in a number of species is normally dispersed; the red-footed cormorant (*Phalacrocorax gaimardi*) of South America, for instance, often nests on cliffs where only scattered sites are available. Similarly, tropic birds nest in holes or crevices in cliffs or under the shelter of rocks or bushes. The sites are thus normally well dispersed, and in some places there is intense competition for them.

Pair formation displays

Courtship behaviour. During courtship (except in tropic birds) the male typically selects a nest site and then displays from this site to passing females. In the male advertising display of boobies ("Sky-pointing") and of cormorants and anhingas ("Wing-waving"), the head and tail are raised and the wings partially raised. The boobies have species-specific positions of the wings and tail; the cormorants and anhingas wave either their wing tips or their whole wings, to a varying degree. In many species a particular call is associated with the display. In the "Rattling" display of frigate birds, the male leans back on his tail, extends his wings along the ground, and with his bill pointing verti-



Male blue-footed booby (*Sula nebouxii*) "Sky-pointing," a pair formation display.
Eric Hosking

cally throbs his inflated scarlet throat (gular) sac, vibrates his wings, and claps his bill.

In most pelecaniforms, when a female has finally joined a male at the nest site, he may go and fetch nest material, which is then added to the nest structure by the female, alone or with the assistance of the male. There is no courtship feeding in the pelecaniforms, but in tropic birds there are a few records of one adult feeding the other during incubation. Copulation normally occurs on the nest site, without any special precopulatory or postcopulatory displays. The pair-bond, once established, is reinforced by such activities as joint nest building and defense, and preening of one bird by the other. Certain displays functioning in individual recognition are performed in the air or on the ground when one of a pair returns to the nest and also occur during the "handling" of nest material, which continues during incubation. The boobies have a "Head-wagging" recognition display; the cormorants and anhingas have a comparable "Kink-throat" display; similarly, the frigate birds use a "Rattling" display.

In tropic birds, which do not normally have nest sites in the open, courtship display is aerial. Birds fly around in a group, calling, then two individuals may leave the group and fly together, the upper bird depressing his tail (especially the long central feathers) during gliding or hovering; the flights often end with one or both birds flying into a nest cavity. Tropic birds do not collect nest material but probably sometimes enlarge nest cavities with their bills or feet.

In all or nearly all pelecaniform families, mating frequently occurs with the same partner in successive years, and in many populations the same nest site may be occupied repeatedly. Since most species nest colonially, fighting over nest sites is common. It is particularly intense in the gannet and in populations of tropic birds when nest sites are in short supply.

Breeding seasons. Outside the tropics, egg laying by pelicans, boobies, and cormorants occurs typically in the spring, and the laying period extends over one to three months. In the tropics there is greater diversity in both the timing and extent of the laying period. Although many tropical populations have distinct laying periods at the same season in each year, a number of those in relatively seasonless environments have been known to lay in every month of the year. The blue-footed booby (*Sula nebouxii*) in the Galápagos, for example, is known to breed at intervals of less than a year. In other populations, laying is synchronized but occurs at nonannual intervals; *e.g.,* the brown booby (*S. leucogaster*) on Ascension Island has laying periods roughly every eight months. For the cormorants (*Phalacrocorax carbo* and *P. africanus*) of Lake Victoria, conditions are favourable for breeding during about eight months in each year, but the scarcity of islands free of predators forces the birds to occupy the few safe colonies in relays.

Nests, eggs, and young. Most pelecaniforms build substantial but untidy nests of twigs, grass, algae, or feathers; in many species guano accumulates on and around the nest as breeding progresses. Many cormorants and pelicans habitually build their nests on the ground, but most species that normally breed where trees or bushes are available build their nests in them. Certain cormorants and boobies sometimes nest on cliff ledges, and tropic birds and most boobies lay their eggs on the open ground in insignificant depressions.

Clutch size

There is much variation in clutch size. In general, species living in the unproductive tropical seas lay only a single egg. Those living in seasonally productive areas at higher latitudes, or in areas where surface waters are enriched by the upwelling of nutrient laden water from deeper layers, lay larger clutches and raise several young at a time. Tropic birds and frigate birds all have clutches of one egg, but pelicans generally lay one to three eggs, cormorants two to five, and anhingas three to five. The boobies show a more complex picture. The gannet of the temperate zones and the tropical red-footed booby (*Sula sula*) have clutches of one egg, but the other typical boobies of the tropics—the masked and brown boobies—normally lay two eggs. These two-egg species, however, normally raise only one chick, the older nestling apparently pushing the younger out of the nest soon after hatching. The second egg in these ground-nesting species seems to function as an insurance against infertility, loss, or breakage of the first, risks that may be lower in the tree-nesting, red-footed booby. The single-egg clutch of the gannet has not been satisfactorily explained, since pairs have been able to raise two chicks when an extra has been provided experimentally.

Except in the tropic birds, the eggs of pelecaniforms are small in relation to body size. In the tropic birds they are 9–13 percent of the adult weight, in the frigate birds, 6 percent, in the boobies, 3–5 percent, and in the pelicans and cormorants, 2–3 percent. In species with clutches of more than one, eggs are laid at intervals as short as 24 hours, or as long as five to six days in the case of the masked booby. In all of the groups, both parents participate roughly equally in incubation and feeding of the young. The lengths of the incubation shifts vary extensively, both within and between species. In some species (*e.g.,* many cormorants) shifts last only a few hours, but in tropic birds and some boobies shifts of three or more days are not uncommon; in the great frigate bird (*Fregata minor*), on the Galápagos, shifts last 10–15 days.

Incubation generally starts with the laying of the first egg (though not in some cormorants). First and last young may hatch several days apart and often differ substantially in size during the nestling period. The incubation period is just under four weeks in anhingas, four to four and a half weeks in cormorants, about four and a half weeks in pelicans, six weeks in tropic birds, and six to eight weeks in frigate birds.

Growth of young birds

The young, which are helpless and essentially naked at hatching, are brooded continuously for several days, until their down grows. The tropic birds are exceptional in that they have thick down at hatching and are sometimes left alone by their parents even on the day of hatching. The rate of development of the young varies dramatically and is correlated with the characteristics of the environments in which the birds live. The fledging period in the pelagic tropical masked booby is double that of cormorants of comparable body size living in the cool, productive waters off the coast of North America (17 weeks versus less than eight weeks). Brown pelicans in North America fledge in less time (nine weeks) than white-tailed tropic birds weighing only one-tenth as much (10 weeks) and in less than half the time needed by tropical frigate birds weighing about one-third as much (more than 20 weeks).

The young of all pelecaniforms are fed on regurgitated food that they obtain by inserting their heads into the mouths of their parents. In frigate birds and tropical boobies, the parents continue to feed their young for a long period after they can fly, apparently because the young take a long time to become sufficiently skillful in hunting to survive in tropical areas where food is generally scarce. The most extreme case is that of the great frigate bird, in which the young can fly at about five months of age but sometimes are fed for more than a year after this. In the tropic birds and the gannet, however, the young are not fed by the parents after fledging; they leave their birthplace either by flying straight out to sea or by fluttering down to the water and swimming out to sea.

Molt. In the pelecaniforms, as in most birds, replacement of the feathers (molt) occurs mainly in the intervals between breeding. In a few species molt progresses concurrently with breeding activities, but normally, a complete molt commences when breeding is over. In the anhingas the primary flight feathers of the wing are shed simultaneously, but in the other groups molt is gradual. The primary and secondary feathers typically are replaced by means of two or three molt waves that start several feathers apart and move outward in parallel until all feathers have been shed. In most, if not all, pelecaniforms, feathers of the head and neck, and some on the body, are replaced a second time shortly before the next breeding season. Special feathers (crests, neck plumes, white patches) acquired at this time are shed around the time of egg laying in some species; in pelicans some head feathers are replaced yet again during incubation.

In many of the pelecaniforms the colours of the soft parts (especially the bare skin on the face) change as the birds become sexually active. In some pelicans a horny triangular plate grows on the upper mandible towards the tip before the breeding season and is shed after laying.

Causes of death

Ecology. *Survival and mortality.* Full-grown pelecaniform birds have few natural enemies, and although some are taken by marine predators they are generally long-lived. Nestlings, recently fledged young, and sometimes even adult birds suffer heavy mortality when food shortages occur. In many tropical areas where food is generally scarce, irregular fluctuations in the supply lead to drastic variations in breeding success. Mortality from starvation also occurs among birds of richer seas, most dramatically along the west coast of South America in years when oceanographic changes cut off the normally abundant food supply of the huge bird populations. Even under average conditions, young pelecaniforms in their first year after fledging experience much higher mortality than adults. In the European shag (*Phalacrocorax aristotelis*) more than half the young die during this period, although among adults annual mortality is only about 15 percent in males and 20 percent in females. In the British population of the gannet, about 80 percent of the fledglings die before reaching breeding age (about five years), most during their first year. Adults die at an average rate of less than 6 percent per year and have a mean expectation of life exceeding 16 years. Age records for individual banded birds include a great frigate bird that was recently found breeding at an age of more than 30 years and a masked booby breeding at about 23 years. Negative evidence suggests, however, that few individuals of these species survive much longer than the two examples.

Population movements. Members of the pelecaniform populations breeding at high latitudes generally move to lower latitudes in winter but do not perform transequatorial migrations. Adequate analyses of banding results have been carried out on only a few species. Great cormorants (*Phalacrocorax carbo*) from Britain do not show a well-defined migration but disperse random distances from the colonies. In all populations of gannets, juveniles move further than adults. A number of juvenile gannets from New Zealand have been found in eastern Australia only a week after leaving their natal colonies on their first flight, having made journeys exceeding 2,600 kilometres (1,600 miles), at average speeds of more than 370 kilometres (230 miles) per day. Gannets that survive to breeding age usually return to the same cluster of nests where they were hatched. In a similar way, juvenile lesser frigate birds (*Fregata ariel*) from colonies in the central equatorial Pacific disperse on a broad front, many moving over 6,500 kilometres (4,000 miles) from their natal island. Their pattern of dispersal may be related to that of the prevailing winds in the area.

Feeding ecology. The various groups of pelecaniform birds are specialized for different ways of life and in particular for different feeding methods. Tropic birds and boobies, though they differ in many ways, are both adapted for catching prey underwater by plunging from the air. They are powerful flyers and buoyant swimmers, but, although their feet have large webs, their legs are not streamlined and they are not specialized for fast swimming. The feet are probably used underwater more for steering than for propulsion, and the birds depend mainly on the impetus of the dive to enable them to approach their prey at high speed.

Tropic birds are adept at hovering while locating prey, and probably obtain most of their food near the surface. Many boobies often dive from greater heights and probably go deeper than tropic birds; the gannet sometimes dives from more than 30 metres. Blue-footed boobies, when hunting in groups, tend to dive almost simultaneously. A disyllabic whistle is often heard from such groups as they start to dive and may be a signal given by the initiator. It has been suggested that the simultaneous plunging of several birds may confuse the fish in a school and so increase each bird's chance of catching one. Red-footed boobies, and perhaps also other boobies, catch flying fish (family Exocoetidae) in the air as well as in the water. These fish, with squid of the family Ommastrephidae, are staple foods of the tropical boobies and the tropic birds. Boobies of upwelling zones and the gannet feed almost entirely on fish.

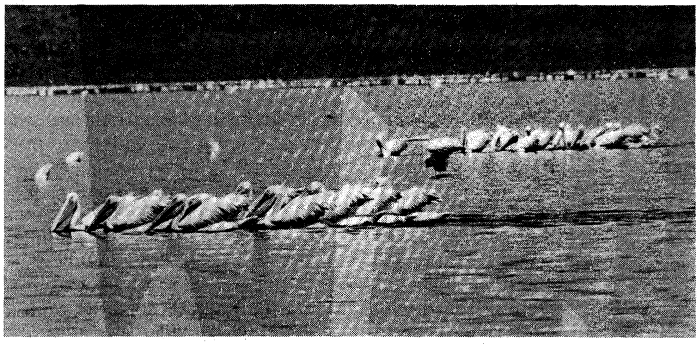
Frigate birds are among the most aerial of all sea birds. Their best known feeding habit involves piracy, in which they harry other sea birds until they disgorge their prey, after which the frigate birds catch the food in the air or pick it from the surface. They also catch prey for themselves, however, pursuing flying fish when they leap through the air and snatching fish (and probably squid) from the surface without alighting.

Feeding by pelicans

Pelicans generally catch their prey while swimming, thrusting their long bills and long necks below the surface to scoop up fish in their distensible throat pouches. When fishing in shallow water, pelicans often cooperate to form a kind of living net. They form a crescent facing the shore, or even a circle, and then close in, splashing, paddling hard. Each bird keeps station until the fish panic and can be captured as they try to escape between the birds. Pelicans have also been observed herding and capturing ducklings in a similar way. The brown pelican, unlike the other species, often forages some distance offshore and habitually fishes by plunging from the air.

Cormorants and anhingas are adapted for underwater swimming. The cormorants pursue free-swimming or bottom-living fish, and some species also eat mollusks and other invertebrate animals. When feeding on schooling fish, cormorants often engage in mass fishing activities in which a flock advances in a long line stretching at right angles to the direction of movement, apparently with the

Leonard Lee Rue III—Bruce Coleman, Inc.



Flocks of white pelicans (*Pelecanus onocrotalus*) fishing cooperatively on Lake Nakuru, Kenya.

fish fleeing ahead of them. The birds swim forward while above the surface and also while pursuing prey underwater, and laggards fly forward and land just ahead of the line. Anhingas do not pursue their prey but lie in wait underwater and then stab passing fish. Most of the fish that they eat are slow swimming and laterally flattened.

The physiological adaptations for diving have been little studied in cormorants and anhingas. It is known, however, that for cormorants feeding on the bottom the mean length of the dives is directly related to water depth. Dives of most species usually last less than half a minute, although dives as long as one minute are not uncommon. The time spent resting after a dive averages between one-half and one-third of the duration of the dive, but if dives are very long the length of the rest periods approaches that of the dives. Anhingas often stay under water for as long as two minutes, and a dive of nearly seven minutes has been recorded in a captive bird.

FORM AND FUNCTION

Distinguishing characteristics. Few generalizations can be made about the gross morphology of the pelecaniforms, since their external form reflects the diversity of their adaptations. In the larger and more aerial species, the skeleton is extensively pneumatized: nearly all the bones contain air sacs that are connected with the respiratory system. In the frigate birds, pelicans, and boobies, the major limb bones are no more than thin-walled hollow tubes with some internal struts. Pneumatization is least developed in anhingas, which need to have a high specific gravity to remain underwater while motionless.

Cushioning by air sacs

Those pelecaniforms that plunge into the water from a height (boobies, tropic birds, and the brown pelican) have a system of air sacs under the skin, connected with the lungs, that form a spongy mattress that presumably protects the bird as it strikes the water, although its presence also in species of pelicans that do not plunge suggests that the buoyancy it confers may be advantageous in itself.

The reduction of the external nostrils in most pelecaniforms is probably also connected with diving and swimming underwater. The closure, complete in boobies and in adult cormorants and anhingas, is only partial in frigate birds and pelicans; the nostrils are developed normally in tropic birds. The reduction of the nostrils in the aerial Fregatae as well as in the highly aquatic Pelecani suggests (as does the structure of the foot) that the Fregatae originated from a more aquatic stock. In the birds with nonfunctional nostrils, air enters the mouth cavity directly from the outside and then reaches the lungs in the normal way via the glottis—*i.e.,* the muscular opening to the windpipe—at the base of the tongue. In boobies and cormorants entry of air to the mouth when the bill is closed is by means of secondary external nostrils. These are slitlike openings at the angle of the mouth on each side where a horny flap (jugal operculum) at the base of the upper mandible overlaps the lower mandible.

Tropic birds, pelicans, and boobies have water-repellent plumage and are very buoyant. They swim high in the water and can reach significant depths only by acquiring momentum during aerial dives; their return to the surface is rapid. Frigate birds have wettable plumage, cannot swim or dive, and quickly become waterlogged if they alight on the water. In cormorants and anhingas the large contour feathers are wettable and trap little air when the birds are swimming underwater; this keeps buoyancy low in these birds and enables them to submerge from the surface without benefit of an aerial dive. The compressible but dense body plumage remains dry near the skin, at least in many species. The hind limbs of anhingas, which are not specialized for fast swimming, are relatively efficient for climbing and perching. Anhingas do not normally take flight from the water but climb out and dry their wings in a characteristic spread-wing posture that is also commonly used by cormorants. Once dry, anhingas show highly manoeuvrable flight and can also soar well, in contrast to the cormorants, which have a higher wing loading. Many species of pelecaniforms habitually spread their wings in "sun-bathing" postures, even when dry. The function of this behaviour is not well understood.

Physiological adaptations. Most pelecaniform birds nest in exposed situations, and special adaptations are necessary to prevent overheating of adults and young, particularly in tropical areas. The masked booby on the Galápagos, for example, stands with its back to the sun, thus shading its feet and naked throat pouch and permitting loss of heat by convection and conduction; it elevates its scapular (shoulder) feathers and droops its wings allowing maximum exposure of shaded feather areas, and it exhibits sustained fluttering of the throat pouch. Throat fluttering, which permits evaporative cooling with minimum expenditure of energy, is used under heat stress by all the pelecaniforms except the tropic birds, which do not have a naked throat pouch. The rate of throat fluttering remains roughly constant under increasing heat stress but becomes continuous instead of intermittent, and there are increases in its amplitude and in the throat area involved. All species make use of panting to increase evaporative cooling. Under heat stress the rate of breathing increases with increasing body temperature, and when it reaches a maximum the amplitude of breathing increases.

Although even the newly hatched young of certain species are capable of throat fluttering, they cannot regulate their temperature when left in the sun, until they have acquired down, and in some species not until they are several weeks old. Young tropic birds provide an exception: they have dense down at hatching and are sometimes left alone for long periods within a few days of hatching; small chicks are capable of regulating their body temperature by panting, even in the sun, but most nests are in the shade.

Salt excretion

Evaporative cooling requires much water, which is obtained from the food and probably also by drinking seawater. The resulting high intake of salt necessitates special adaptations for elimination of the excess. Marine birds have highly developed salt excretory glands, the lateral nasal glands, which function when the amount of salt taken into the body is more than the kidneys can deal with. The glands secrete a concentrated solution of sodium chloride, which flows through ducts to the nasal cavity. In most sea birds, including the tropic birds, it then flows out through the nostrils and may drip or be shaken off the tip of the bill. In boobies and cormorants, in which the nostrils are closed, the solution trickles from the internal nares in the roof of the mouth and out at the end of the bill. In some species of cormorants that occur in both marine and freshwater habitats, the salt gland is relatively larger in individuals from marine habitats, presumably reflecting higher salt intake.

CLASSIFICATION

Distinguishing taxonomic features. The classification of the order Pelecaniformes is based mainly on morphological features. Recent attempts have been made to clarify relationships by analysis of egg-white proteins and by comparative analysis of social behaviour. The parasitic feather lice (Mallophaga) found on members of the different groups also provide clues to relationships.

The six modern families of pelecaniform birds are readily distinguished, although the relationships among them are not firmly established. The families are internaly homogeneous, and in each of them the modern species are often placed in single genera; the Phaethontidae, Fregatidae, Pelecanidae, and Anhingidae are usually treated in this way. The relationships of many fossil groups—indicated below by a dagger (†)—are quite uncertain.

Annotated classification.

ORDER PELECANIFORMES

Aquatic birds with totipalmate feet (webs connecting all 4 toes); 1st toe (hallux) not pointing backward as in most birds but turned inward and joined by a web to 2nd toe.

Suborder Phaethontes

Family Phaethontidae (tropic birds)

Medium sized (16–19 in. or 40–48 cm long). Plumage satiny white with black markings; central 2 tail feathers much elongated in adult. Sexes similar. Bill laterally compressed, slightly decurved, and with undivided horny sheath (rhamphotheca). Nostrils open in slits; nasal glands each with 2 ducts. No naked gular (throat) pouch. Neck short, 14–15 cervical vertebrae. Pelvic girdle reduced, legs short, set far back. No nest-building behaviour. Single egg heavily marked with brown and black. Chick hatched covered with grey down. 1 fossil genus, *Prophaeton* (lower Eocene of England); 1 extant genus, *Phaethon;* 3 living species, worldwide in tropical oceans.

Suborder Fregatae

Family Fregatidae (frigate birds)

Large (31–34 in. or 79–86 cm long), light bodied. Plumage blackish, juveniles and some adults with white underparts; juveniles with white heads, sometimes tinged orange. Wings very long, pointed; tail deeply forked. Sexes differ in size (female larger) and in coloration of plumage and soft parts. Bill long, slender, strongly hooked at tip; horny sheath divided by deep grooves. Nostrils almost closed; nasal glands each with single duct. Small gular pouch, highly distensible and brilliant red in courting male. Neck short, 14–15 cervical vertebrae. Furculum (wishbone) fused with coracoids and with sternum (breastbone). Pelvic girdle and leg muscles reduced; legs very short. Feet small, webbing only between bases of toes. One white, thin-shelled egg. Young naked at birth, soon acquire thick white down. No fossil record. One extant genus, *Fregata;* 5 living species, worldwide in tropical oceans.

Suborder Pelecani (pelicans, boobies, cormorants, anhingas)

Horny sheath of bill divided by deep grooves (largely suppressed in anhingas). Naked gular pouch; facial skin and pouch colourful, especially in breeding season. Nasal glands each with a single duct. Neck long, with 17–20 cervical (neck) vertebrae; 8th and 9th have articulations permitting sharp S-bend, especially in anhingas. Pelvic girdle not reduced, legs strong. Feet large, fully webbed. Eggs thick shelled, with chalky white outer layer. Young naked at hatching.

Family Pelecanidae (pelicans)

Very large (50–72 in. or 127–180 cm long). Plumage white, brown, or grey; some species crested. Wings long, broad; tail short. Sexes similar. Bill long, hooked at tip, with enormous gular pouch suspended from highly distensible rami (lateral elements) of lower mandible. Nostrils almost closed. Furculum fused with sternum. Down of nestlings white. 10 fossil species in the genera *Pelecanus* (the earliest from lower Miocene of France) and *Liptornis* (middle Miocene). One extant genus, *Pelecanus;* 7 living species, seacoasts and freshwater lakes of southern Africa, central and eastern Eurasia, Australia, and North America; north and west coasts of South America.

Family Sulidae (boobies)

Large (26–40 in. or 66–102 cm long). Plumage of most species white with black on wings; some brown and white; juveniles largely grey or grey-brown. Wings long and narrow, tail rather long, wedge shaped. Sexes generally differ in voice, sometimes in colour of bare skin. Bill stout, conical, slightly decurved near the tip and finely serrated on cutting edges. Nostrils closed. Down of nestlings white. 18 fossil species, in the genera *Sula* (the earliest *S. ronzoni,* lower Oligocene of France), *Microsula, Morus, Palaeosula* and *Miosula.* Two extant genera, *Sula* with 6 living species, *Morus* with 1 species (sometimes considered 3 species, forming a superspecies); widespread but irregular distribution, on seacoasts and oceanic islands.

Family Phalacrocoracidae (cormorants)

Medium to large (19–40 in. or 48–102 cm long). Plumage usually blackish; some with white underparts, and a few largely grey; some species have ornamental crests or patches of white when breeding. Wings broad, tail long and stiff. Sexes similar, but females up to 30 percent lighter in weight. Bill slender, sharply hooked at tip. Nostrils closed in adults. Bony occipital style projecting upward from back of skull. Down of nestlings dark. 24 fossil species in the genera *Graculavus* (the earliest, upper Paleocene of North America), *Actiornis, Pliocarbo,* and *Phalacrocorax.* About 28 living species (and one recently extinct), all currently included in the genus *Phalacrocorax;* certain Old World species have sometimes been placed in a separate genus *Halietor,* and the Galápagos flightless cormorant (*Phalacrocorax harrisi*) in the genus *Nannopterum.*

Family Anhingidae (anhingas or darters)

Large (34–36 in. or 86–91 cm long). Plumage mainly black and brown with pale markings on back and wings; breeding male has light-coloured plumes on head and neck. Tail very long and stiff. In some forms the sexes differ in colour. Bill long, straight, slender, sharply pointed; cutting edges serrated towards tip. Nostrils closed in adults. Occipital style smaller than in cormorants. Head small, neck long and slender. Legs short but strong. Pyloric lobe of stomach with mat of hairlike processes. Down of nestlings brown. Five fossil species, in the genera *Protoplotus* (the earliest from the Eocene of Sumatra), and *Anhinga.* One extant genus, *Anhinga;* living forms have been treated as 1, 2, or 4 species and are here considered as 4 species comprising a single superspecies.

†*Family Elopterygidae*

Fossil. Marine birds probably related to the Pelecani, perhaps especially to the cormorants. Genera *Elopteryx* (Upper Cretaceous, Maestrichtian, of Romania); *Argillornis* and *Eostega* (Eocene of Europe).

†*Family Pelagornithidae*

Fossil. Marine birds probably related to the Pelecani, perhaps especially to boobies, but not adequately known. Genus *Pelagornis* (middle Miocene of France).

†Suborder Odontopteryges

Fossil. Very large marine birds, with long wings and short legs. Bill long and high, with sharply pointed, toothlike projections on margins of upper and lower mandibles, continuous with bony structure of jaws (and thus not true teeth). Upper and lower mandibles with longitudinal grooves, suggesting that their horny sheaths were divided.

†*Family Odontopterygidae*

Fossil. "Teeth" slant forward. Nostrils closed or nearly so. About half the size of the pseudodontorns (see below). Genus *Odontopteryx* (lower Eocene of England).

†*Family Pseudodontornithidae*

Fossil. "Teeth" vertical—not slanted forward. Wingspread about 4.5 metres. Genera *Pseudodontornis* (Miocene); and *Osteodontornis* (upper Miocene of California), one of the best preserved of early marine birds.

†*Family Cyphornithidae*

Fossil. Based on fragmentary material and with uncertain relationships but may be members of this suborder. Lower Miocene, North America. Genera *Cyphornis,* larger than pseudodontorns; *Palaeochenoides,* larger than pelicans, may be synonymous with (and take priority over) *Pseudodontornis;* *Tympanonesiotes,* smaller than the preceding form but may be related.

†Suborder Cladornithes

Fossil. Relationship to pelecaniforms doubtful. One family, with 1 genus, *Cladornis* (lower Oligocene of Patagonia); known only from a poorly preserved tarsometatarsus.

Critical appraisal. The tropic birds (Phaethontes) differ in many ways from the other members of the order Pelecaniformes and may well be wrongly placed in this order. They are united to the others primarily by the totipalmate condition, but the first toe in Phaethontes is distinctly more elevated on the tarsometatarsus than in the other groups. Some of the resemblances of the Phaethontes to the Laridae (gulls and terns) in the order Charadriiformes may not result from evolutionary convergence but may indicate that the Phaethontes arose from forms more closely related to the basal stock of the Laridae than to that of the Pelecaniformes.

The frigate birds (Fregatae) are generally treated as a distinct suborder, on the basis of a number of morphological characters that they share with members of the order Procellariiformes but not with the Pelecani. Pending further evidence from protein structure, this distinction seems worth maintaining.

There is little doubt that the suborder Pelecani is made up of closely related groups. The pelicans are sometimes placed in a separate superfamily, but it is more useful to emphasize the similarities among the modern groups by keeping them together. The anhingas have sometimes been treated as a subfamily of the Phalacrocoracidae, but their morphological and ecological differences from cormorants perhaps justify separation at the family level.

Odontopteryges show a number of similarities to Procellariiformes as well as to Pelecaniformes and also show differences from both. The suggestion that they deserve to rank as a separate order Odontopterygiformes will be strengthened if it can be proved that *Osteodontornis* lacked or had a much reduced first toe, since this is an essential element of the totipalmate foot of pelecaniforms.

Available evidence suggests that the stocks leading to the Procellariiformes, Odontopteryges, typical pelecaniform birds (Fregatae plus Pelecani), Phaethontes, and Laridae, diverged during the Cretaceous Period, probably at least 70,000,000 years ago. (N.P.As.)

Anseriformes (screamers, waterfowl)

The order Anseriformes includes the well-known ducks, geese, and swans (family Anatidae) and the little-known screamers (family Anhimidae). The Anatidae comprise about 142 species of medium to large birds, usually associated with aquatic or marine habitats and known collectively as waterfowl (in the United States) or wildfowl (in Europe). These birds are of interest to many people for varied reasons: to the romantic and the artist for their swift, massed flights and association with wild places; to the sport hunter as a difficult and worthwhile quarry; to the economist for exploitation as producers of meat and feathers; to the farmer as possible competitors for grass or as damagers of crops; to the medical worker as a possible reservoir of arthropod-borne viruses. The conservationist views them as biological indicators of the richness of wetland habitats and, like those habitats, as objects for his urgent attention. To the zoologist they are a close-knit group in evolution, but one that has adapted into a wide variety of ecological niches.

Drawing by D. Malick

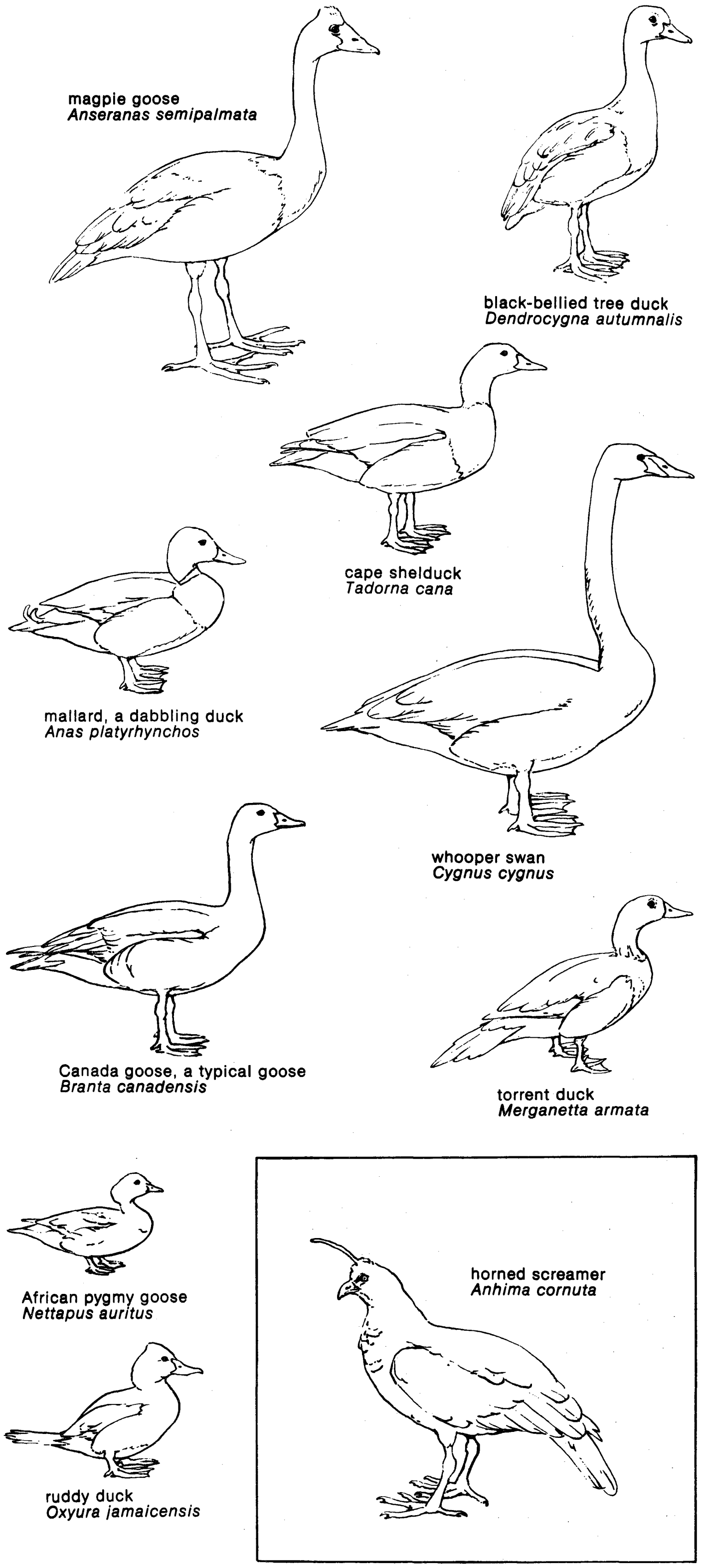


Figure 17: Comparative body plans of anseriform birds.

GENERAL FEATURES

Members of the Anatidae vary greatly in size, the largest being some 60 times as heavy as the smallest. Most are rather chunky birds (see Figure 17), with lengths from just under 30 centimetres to slightly over 150 centimetres and weights from about 300 grams to more than 18 kilograms. The neck is medium to long. The bill also varies widely, as befits birds with diets as different as grass, fish, and minute plankton, but is of medium length, typically broad and somewhat flattened, with a rounded tip, often with a pronounced nail, giving a slight hook to the tip. The bill is often provided with fine lamellae, short parallel plates at the edges of the mandibles that aid in food handling and, in some species, in straining tiny food organisms from water or mud. Anseriform behaviour patterns are an intriguing mixture of the innate and the learned; some so fixed that they can be used as taxonomic characters, others so pliable that completely new traditions can evolve within a few generations.

The three species of screamers are quite different from waterfowl in general appearance. They are moderately long-legged birds about the size of a turkey, with lengths from about 70 to 95 centimetres and weights of 2–5 kilograms. The beak is rather chicken-like, with a pronounced downward hook. The feet are exceptionally large, the toes webbed only at their bases.

Distribution

Anseriform birds breed on every major continent and island except the Antarctic, and many undertake vast migrations over thousands of kilometres. Some species are widespread and number in the millions; other are represented by a few hundred confined to a single island. In the former category come the pintail (*Anas acuta*) and the mallard (*Anas platyrhynchos*), both found throughout the Northern Hemisphere; in the latter, the Hawaiian goose (*Branta sandvicensis*) and the Madagascar white-eye (*Aythya innotata*). Extinction has overcome four species within the last century, and island forms, such as the endangered Marianas mallard, are very vulnerable. The Hawaiian goose was once down to fewer than 50 birds, the Laysan race of the mallard to seven individuals. The creation of wetland refuges has helped to restore these and other declining populations. Nonetheless, even the more common waterfowl, which teem in thousands at certain times and places, are not secure from drastic declines in numbers.

IMPORTANCE TO MAN

The large size of waterfowl has made them an inevitable quarry for man from prehistoric times. Not only do they provide protein but also large amounts of fat and feathers, especially down, much prized in the colder regions. Among the unusual uses of waterfowl parts may be mentioned the conversion of swan tracheae into children's whistles in Lapland, and the eating of the billknob of the king eider (*Somateria spectabilis*) as an aphrodisiac in Greenland. Wary and difficult to approach in their watery haunts, waterfowl exercised man's ingenuity to kill them before the advent of efficient weapons. The period of flightlessness was early discovered and exploited, the birds being driven into corrals of stone or netting. From the latter evolved the Dutch method of catching full-winged ducks by enticing them up large net-covered pipes leading from the secluded pond of a duck-decoy. The birds were seduced into the pipes by food or by their predator-following reaction to a trained dog. Many other ingenious traps were devised, from the clapnets of the ancient Egyptians to the rocket-propelled nets of today's research workers.

Threatened species

Anseriform populations are now so threatened by the loss of their essential wetland habitats to drainage and technological developments that they are no longer able to withstand commercial exploitation. This is generally recognized in civilized countries, where market hunting is banned and efforts are made to control sport hunting so that annual kill does not exceed annual production. As waterfowl migrations pay no heed to national boundaries, the need for international agreement on their conservation is paramount.

As the amount of human leisure increases, so does the importance of nondestructive recreational use of waterfowl. A great flock of waterfowl is one of the remaining spectacles of nature and one that can be made accessible to crowds of people led through covered approaches to concealing blinds. The refuges created to hold reserves of stocks for shooting will, more and more, be developed with the spectator in mind. Such refuges must be numerous and large enough to give adequate feeding grounds. Extra food, however, can be provided as standing crops or distributed from stores. In these ways the occurrence of damage to agricultural interests in the neighbourhood can be avoided. Complaints of such damage are frequently exaggerated, for in many countries migratory waterfowl arrive after the harvests have been gathered and take only spilled or rejected grain and tubers, performing a useful cleaning operation. Even when geese graze on growing cereals in winter, the subsequent crop yield is not reduced. Conflicts can arise, but these can be resolved by changes in husbandry and by techniques of scaring the birds from the fields without injuring them.

Another aesthetic exploitation of waterfowl has been possible because they readily adapt to captivity. Spending much time walking, swimming, or diving, they are less inconvenienced than most other birds by being pinioned (the tip of one wing removed) to prevent flight. They can then be kept in open paddocks instead of being confined in cages. Zoos and private aviculturists maintain waterfowl collections of varying size, the largest being that of the Wildfowl Trust, Slimbridge, England. Here no fewer than 131 of the 142 existing species have been maintained, and more than 100 have bred. Such a remarkably complete comparative collection has, of course, been useful for educational and research purposes. Breeding in captivity also enhances the possibility of restoring wild populations that have been reduced by overshooting or other excesses. The most notable case has been that of the Hawaiian goose. In 1950 fewer than 50 of these birds existed. Twenty years later, thanks to avicultural efforts at Slimbridge and on Hawaii itself, there were more than a thousand, several hundred of which had been liberated on the original habitat. For a different goal, but also effective, have been the efforts of hunting organizations to breed replacements for birds shot. Several species of geese and ducks have responded well to manipulation of habitat and the provision of nest sites. Artificially encouraged breeding may well become vital as the hitherto untouched vastnesses of northern marshes and tundras, where so many waterfowl breed, are opened up, destroyed, and polluted.

Domestication of waterfowl

In view of their ready adaptation to captivity, it is surprising that few waterfowl have been domesticated. The mallard was exploited 2,000 years ago in China, and 17 varieties have been developed, according to whether meat or egg production was important. The muscovy duck (*Cairina moschata*) was domesticated in Colombia and Peru before the arrival of the conquistadores. The greylag goose (*Anser anser*) has been domesticated for at least 4,000 years. Egyptian frescoes of that age already show changes in shape from the natural form, and eight main varieties are now known. The swan goose (*Anser cygnoides*) of eastern Asia has also been domesticated, with three varieties. Other species, such as the Canada goose, mute swan, and Egyptian goose (*Alopochen aegyptiacus*), have been kept in semidomestication for ease of exploitation but without intensive breeding to change their forms. A remarkable form of exploitation has been that of the common eider. Its breeding colonies in the Arctic and subarctic are protected and concentrated by the provision of nest sites and other techniques. The down with which the female lines her nest and covers the eggs is systematically collected during and after incubation without disturbance or loss of productivity, for these birds are very tame.

NATURAL HISTORY

Life history. Young waterfowl are hatched with a complete covering of down and can take to the water as soon as they leave the nest, within 24 to 48 hours. From the first they forage for themselves, but at least one parent remains with them, guarding, guiding, and, initially, brooding them at night and during inclement weather.

Growth and molt

The downy plumage is retained for from two weeks (small ducks) to six weeks (large swans) and is then replaced gradually by the feathers of the juvenile plumage. The flight feathers, the last to develop, likewise vary in rapidity of growth, taking from five weeks to as many months. At fledging, young ducks are freed from parental care and must make their own way on migration. In species breeding in the far north, this begins soon after fledging, in early autumn. Young geese and swans remain with their parents during their first winter, migrating to and from the wintering grounds in their company. In most large species the juvenile plumage is retained through much of the first year of life. Ducks, however, begin to lose the juvenile body feathers almost at once. Some replace the juvenile plumage with an immature nonbreeding (or "basic") plumage, acquiring the first nuptial (or "alternate") plumage in the second autumn. Other species molt directly from juvenile to nuptial and are practically indistinguishable from adults in plumage and size at the age of six months. Swans and geese do not reach full size until the end of their second year, and even at that age swans still retain some immature feathers. In migratory species social flocking and pair formation occur on the wintering grounds, followed by the return migration to the breeding grounds. In the high-Arctic regions the birds arrive ready to nest as soon as the snow cover goes and the water opens; further south the process is more leisurely. When the female duck begins to incubate, the male generally deserts her and joins forces with other males, often after making a quite extensive molt-migration, a movement to another area some distance from the breeding site. The nuptial plumage is lost, and a dull "eclipse" plumage, rather femalelike, is assumed before the simultaneous molt of the flight feathers. The resulting flightless condition lasts three or four weeks, during which the birds skulk in thick cover or remain on large bodies of water. In female ducks this wing molt is delayed until after the young have fledged. The eclipse plumage is succeeded, in three to six months, by the next nuptial plumage. In the case of swans and geese the male remains and molts with the female and family; immature birds and unsuccessful breeders may, however, make a molt-migration to a separate area.

The foregoing life cycles are typical of waterfowl breeding in north temperate or Arctic regions. In the tropics and south temperate regions, migration of waterfowl may not be necessary. When it does occur, migration may be in response to less regular ecological factors such as the onset of the rainy season. Males and females tend to be similar in plumage and to lack eclipse plumages. The birds are ready to breed at any time that conditions are favourable. Undoubtedly there is a relationship between hormonal balance and the appearance or absence of eclipse or brightly coloured plumages, but the exact mechanism is not yet known.

Life spans

The mature bird has a potentially long life, but this is seldom achieved in the wild. In captivity ducks have lived for 20 years, geese and swans for more than 30; there are reports of geese exceeding 40 years of age. Wild populations suffer heavy losses (up to 70 percent) among first-year birds. The adults then experience annual mortalities of from 10 percent to 50 percent, depending on the extent to which the population is shot and on natural factors. At the higher end of this scale the population is almost completely replaced within three years. With luck and cunning an occasional bird may survive for 15 to 20 years or more.

Behaviour. Maintenance of the plumage in an effective waterproof condition occupies much of the time not spent feeding or sleeping. The bill is used both to stimulate the oil gland (situated above the tail) and to spread the oil. Rubbing the chin and throat on oiled areas also helps the process. Preening occurs at the same time, the fine structure of the feathers being nibbled into the interlocking position necessary to prevent the entry of water. Rearrangement of the feathers involves preening, scratching with the feet, and a general body shake, produced by a muscular contraction sweeping from tail to neck. Various wing stretching movements settle the flight feathers. Bathing movements include dipping the head, beating the wings on the surface and, at high intensity, actual diving or somersaulting through the water. Sleep often follows such maintenance activities, the bill being turned and placed under the scapular (shoulder) feathers. Bathing is often a communal activity, and mutual preening is seen in several species.

Display behaviour

Social displays are one of the most interesting aspects of waterfowl behaviour, and many of the signal movements involved in the displays are clearly derived from maintenance activities. Thus preening dorsally, on the breast, and, especially, behind the wing can be seen in ritualized form in social situations. Likewise, the wing stretch and the general body shake occur in threat or sexual displays.

Most waterfowl are gregarious and have well-developed social integration signals. Many indicate the intention to fly by head-shaking or chin-lifting, so that the group takes off together. In swans and geese vocalization plays a part in this situation and also in group flight. Formal aggressive displays are important in gregarious species to prevent actual fighting and injury and to establish instead a stable dominance order. In threat, the behaviour often tends to increase the apparent body size, as when swans raise their wings and ruffle the scapular feathers. Wing flaps and flicking of the folded wings are common in geese. The chin may be lifted during aggressive display, but more frequently the head is thrust forward, often with the bill open. The striking nature of these gestures is often emphasized by patches of colour, which may be concealed before the movement is made. Since these gestures may occur without associated feather patterns, this suggests that the behaviour patterns evolved first.

Pair-forming displays

The pair-forming displays are well developed and characteristic of the species, which is necessary if mating with closely related and sympatric (coexistent) species is to be avoided. Swans and geese (tribe Anserini) cement the pair bond by the "triumph ceremony," with mutual head waving and calling, typically when the male has driven off an intruder. Male sheldgeese (tribe Tadornini) have a puffing, strutting display. Their females incite them to attack other birds by sideways jabbing movements of the bill. Female incitement behaviour is found throughout the rest of the family, but the male pairing behaviour is more varied. It is particularly striking in the dabbling ducks (Anatini), where it is often social, a group of males displaying around a solitary female, who appears to do the mate selection. Ethologists have endeavoured to give objective names to the various postures and sequences (Figure 18). Thus "head-up-tail-up" involves a simultaneous upward jerk of head and tail and a lifting of the folded wings to display the speculum, a set of metallic coloured secondary flight feathers of the upper wing. The "grunt-whistle" involves throwing an arc of water at the female by a sideways flick of the bill, followed by a rearing up of the body, shaking of the head and tail, and, during the whole sequence, giving the call indicated by the behavioural term. The sea ducks (Mergini) also have elaborate male displays, bowing while producing dovelike coos; flagging the head from side to side; jerking the head back on the tail and kicking up a spurt of water (the "head-throw-kick" of the goldeneye, (*Bucephala clangula*). Perhaps the most bizarre display is found in the stifftails (Oxyurini), such as the ruddy duck (*Oxyura jamaicensis*) of North America. The males cock their tails over their backs, inflate their lower necks, and beat on them with their bright blue bills, producing a chittering sound that terminates in a burp.

Displays that maintain the pair bond are less varied and tend to be more marked in the long-lived geese and swans, which pair for life, than in the short-lived ducks. Mutual preening and drinking displays are of this category, as are precopulatory displays, often ritualized feeding movements such as mutual head dipping, bill dipping, or

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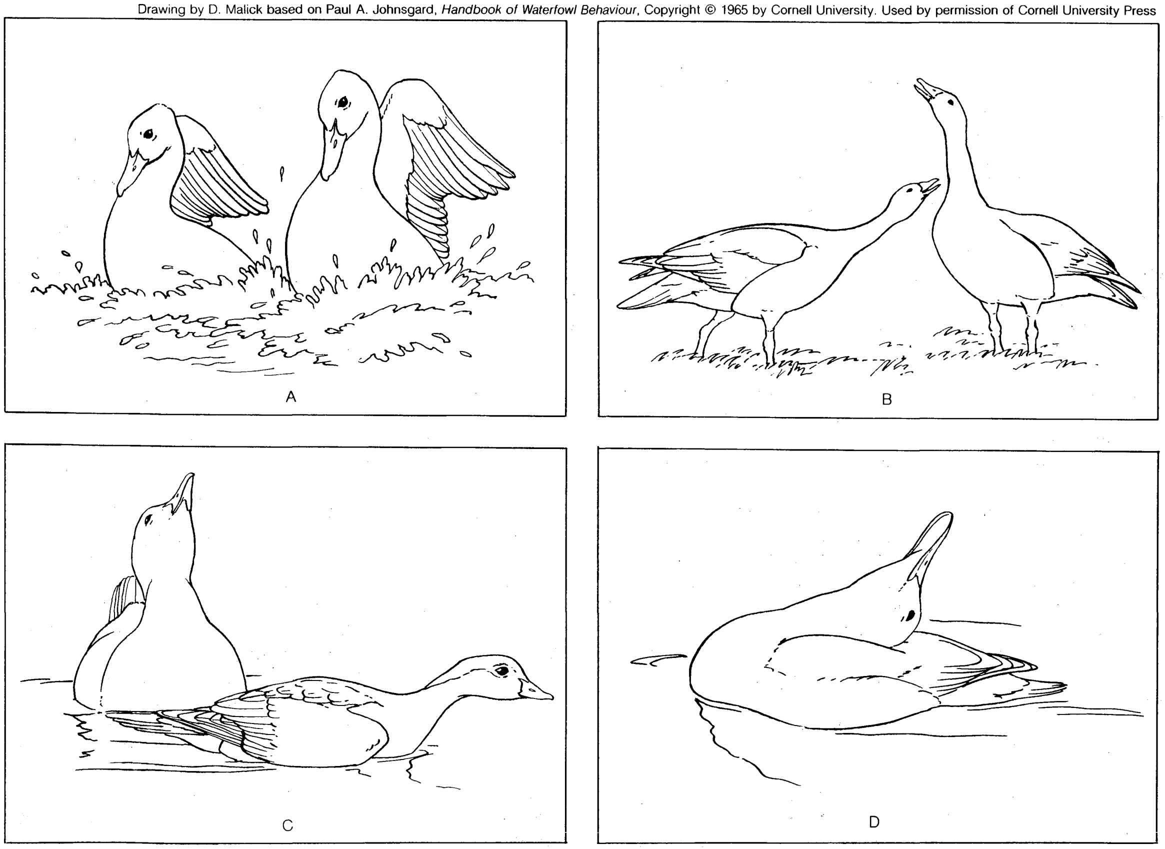


Figure 18: *Waterfowl behaviour.*
(A) Postcopulatory joint display of the fulvous tree duck (*Dendrocygna bicolor*). (B) Triumph ceremony of the bar-headed goose (*Anser indicus*). (C) Female copulatory position of the wood duck (*Aix sponsa*). (D) Head throw display by male canvasback (*Aythya vallisneria*).

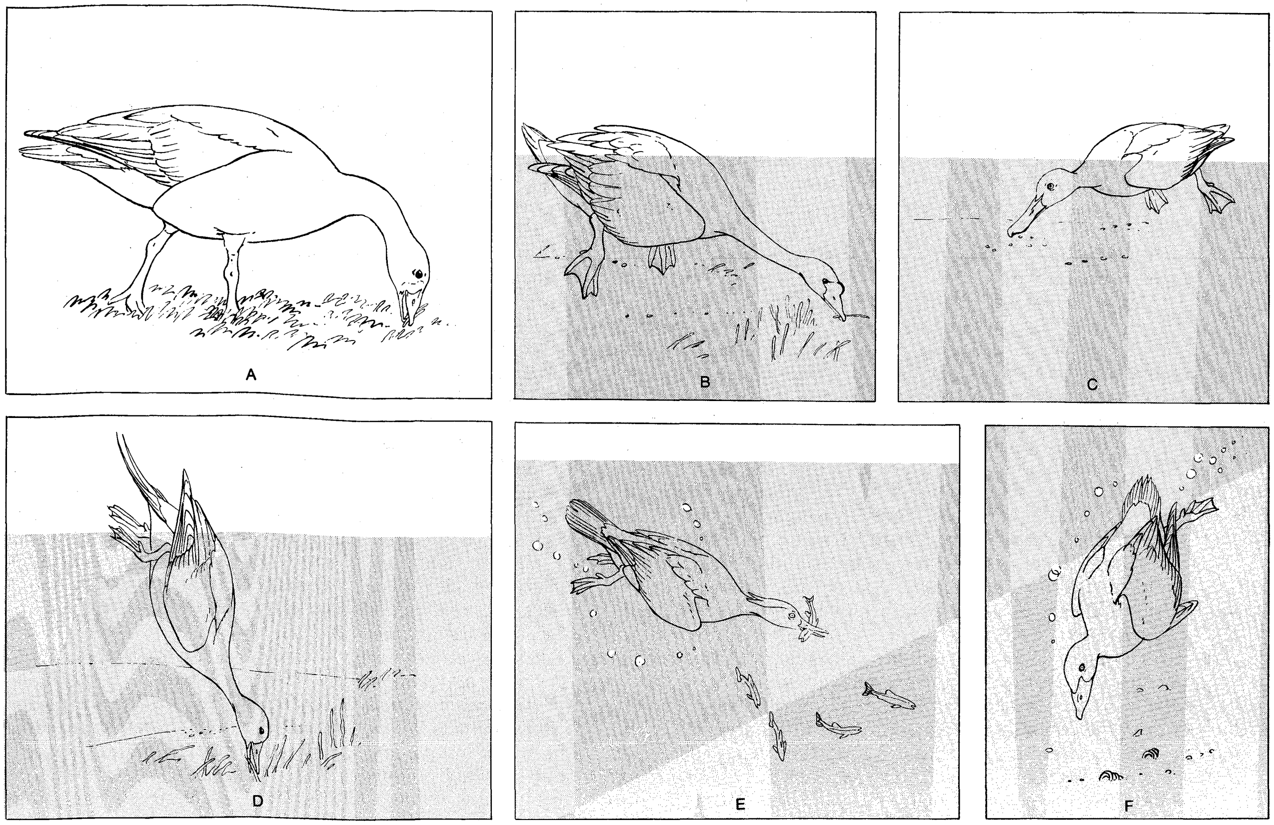


Figure 19: *Feeding characteristics of Anseriformes.*
(A) Grazing; (B) cropping underwater vegetation; (C) filter feeding; (D) dabbling; (E) diving for fish; (F) diving for bottom invertebrates.
Drawing by D. Malick

head pumping. Following copulation, other mutual displays occur. The geese and swans indulge in calling with upstretched necks and lifted wings. More extraordinary is the "step-dance" of the whistling ducks (Dendrocygnini), in which the partners rear up side by side, treading water and each raising the outside wing.

Imprinting

The display patterns, whatever their origin, have been incorporated in the genetic code, and they appear to be ready-formed and complete. Although much waterfowl behaviour can thus be said to be innate, other activities are learned during the life of the individual. One special type of learning, which is called imprinting, has been much studied in this group. The newly hatched, downy young exhibit a strong but unspecific following response to a large moving object, especially if it emits rhythmical sounds. While following such an object they become "imprinted" with the knowledge of its characteristics and thereafter treat it as the parent, even though it may be wildly unsuitable, a balloon for instance. The class of object is not necessarily treated as the sexual partner when the duck matures, but it does appear that male dabbling ducks learn the plumage patterns of the female from the ducks with which they are reared. The females, however, apparently have the ability to ignore such early experience, and to select the appropriate male when they mature; the male plumage patterns are probably sufficiently simple that their recognition is coded into the female's genetic inheritance.

Many other aspects of waterfowl behaviour show a similar interweaving of innate and learned responses. Thus the young bird hatches with distinct preferences as regards the colour, size, and shape of objects at which it will peck. Within these limits, and those imposed by its bill structure and the habitat to which it is taken by its parent, the young bird can be quite adventurous in its feeding, gradually excluding unpalatable or unsuitable objects.

Feeding behaviour

Three main lines of feeding behaviour have evolved in the waterfowl—diving, dabbling, and grazing. Those that dive for food fall into two groups: inland species (Aythyini) favouring relatively shallow lakes up to six metres deep and predominantly feeding on plants such as pondweeds; and sea ducks (Mergini), using deeper marine waters and feeding on invertebrates and fish. Dabbling ducks feed on the surface or by upending to reach the food on the bottom, largely plant material. The grazing geese and swans take grasses and sedges on dry land or in marshes and grub for roots both on land and under the surface. Many of the geese have adapted to the various farm crops, new feeding traditions spreading rapidly as the birds of a group learn from one another.

Migratory behaviour is likewise partly innate, partly learned. Young ducks of migratory populations set off in the autumn in approximately the right direction, even if there are no old birds to guide them. Others show a fixed orientation, after capture and release, that has no apparent function and so has been termed "nonsense" orientation. In both cases the direction is innate, and so is the ability to determine it with reference to the sun's position (making allowance for its diurnal movement) or to the patterns of the stars. It appears, however, that the details of the sun movement and star patterns have to be learned, as do the characteristics of the home from which the bird starts and the winter quarters in which it ends. Like other migratory birds, waterfowl have the ability to return to the nesting area when they are displaced from it, but it is not clear how far they can be truly said to navigate; *i.e.,* to pinpoint their position in two coordinates, again perhaps with reference to sun or stars. In geese and swans, where the young travel with their parents, the possibility exists of their learning the chains of lakes, rivers, and other landmarks down which they fly and so of "map reading" on subsequent migrations. Certainly the migratory behaviour is highly modifiable in these longer-lived birds. New wintering resorts are adopted if they become favourable because of food supply or lack of disturbance. The migratory urge can be overcome completely in geese that are reared artificially. If the climate is not too severe and food and mates are available, such birds will remain where they grow up, even though they are left full-winged and free to come and go.

One form of waterfowl behaviour that seems to be largely innate is sound production. Vocalizations are rather simple and do not apparently offer the opportunity for individual elaboration found in songbirds. A range of honks, whistles, grunts, coos, and quacks are produced, mainly in the context of social contact and flock cohesion or in courtship display. The voices of the sexes are generally quite dissimilar. Some nonvocal sounds are also produced by the wings of some species, by inflatable air sacs of others.

Locomotion. Only the screamers (Anhimidae), with their broad wings and light bodies, are able freely to indulge in soaring flight. The magpie goose (*Anseranas semipalmata*) has somewhat similar wings, but typically the waterfowl have strong, rapid flight, the heavy body and relatively small wings giving a very high wing loading. The trumpeter swan (Cygnus cygnus buccinator) is probably the heaviest flying bird there is, weights up to 18 kilograms (38 pounds) having been claimed. Trumpeter swans beat their wings about three times a second, the smaller ducks twice as fast. Geese tend to fly in long extended lines, often adopting a V formation. Cruising air speeds of most waterfowl appear to be around 65 kilometres (40 miles) per hour; when pressed they can certainly reach 100 kph. On migration most flocks fly at altitudes of between 300 and 600 metres. Uncommonly they may be seen around 3,000 metres, and the barheaded goose (*Anser indicus*), breeding in Tibet and wintering in India, must fly at 6,000 metres to get through the Himalayan passes.

Flight speeds and altitudes

Swans and the heavier geese require a running start into the wind when taking off from water or land. The Aythyini and the Mergini also skitter along the surface before becoming airborne. Three of the steamer ducks (*Tachyeres*) have lost the power of flight and when in a hurry can only churn along the surface like paddle steamers. Small geese and most ducks can launch themselves directly into the air, pushing with the legs and giving a tremendous thrust with the wings, which may strike the water. In landing, the wings and spread feet are used for braking and guidance. Height is often lost rapidly by alternate sideslips, a process carried to extremes in the downward "whiffling" of geese, in which the birds may actually roll over on their backs, still keeping the head straight and level. The Anatini are quite manoeuvrable near the ground, back-flapping their way vertically down for the last few feet. The Aythyini land faster and normally only on water.

Walking on land is well developed in the longer-legged geese and in gooselike species. The "goose-step," with exaggeratedly lifted feet, is well shown by the spur-winged goose (*Plectropterus gambensis*). Others walk more straightforwardly and when pressed, if flightless for instance, can outrun a pursuing man. In the ducks, whose short legs are situated laterally and posteriorly, the gait is at best a waddle. Most of the Aythyini and Mergini have legs so far back that the body is carried almost upright when out of the water. This development reaches its extreme in the Oxyurini, which rarely leave the water on foot.

The magpie goose, with long, nearly web-free anterior toes and a long posterior hallux, is able to perch on twigs at the tops of trees. But it is very much of an exception. A number of other ducks, especially the hole nesters, can perch on branches or scramble up rough surfaces.

Underwater locomotion

It is for swimming that most of the waterfowl are especially adapted, with their waterproof feathering, fat-insulated body, and powerful legs with webbed feet. The feet paddle alternately in slow swimming, the whole leg being used when the bird is moving fast. All waterfowl will dive if pressed, and about two-fifths use diving as their normal feeding procedure. These submerge by arching the body and propelling themselves forward with both feet so that entry is in a smooth arc. The whistling ducks, however, take a jumping "header," clearing the water like coots, while those consummate divers, the stifftails, simply slip under. Diving in fast-running streams requires great power and adaptations, such as the streamlined form and the long, stiff tail of the torrent ducks (*Merganetta*). Once a diving duck is under water, the legs are sculled together. In some species the wings are opened and used as steering rather than as propelling devices.

Long-tailed, or old squaw, ducks (*Clangula hyemalis*) have been caught in fishing nets as deep as 55 metres, but this is exceptional and most species do not dive much below 6 metres. They normally remain below for less than 30 seconds, occasionally up to 90 seconds, but are physiologically capable of much longer submergence.

Reproductive behaviour. Copulation in waterfowl typically occurs in the water, the pair being in isolation. Indeed, the avoidance of disturbance at this time may be a major function of territorial behaviour. Colonial nesting is not normal, but high densities may occur when certain areas are unusually safe and suitable, islands for instance. More than 200 gadwall (*Anas strepera*) nests per acre (494 per hectare) have been reported in such circumstances. More usually the nests are scattered, the distance between them being determined in part by the strength of the territorial activity. Some swans, shelducks, and sheldgeese are highly aggressive and defend a territory around the nest. In most species the situation is less clear-cut. The pairs of duck remain within a "home range," incorporating water, a loafing spot, nesting cover, and food, but do not vigorously defend its borders. Rather, the territory proper seems to be an area about the female, moving as she does. A male intruding on a pair gives rise to the "three-bird chase," which is a characteristic of prenesting behaviour. Geese likewise defend an area around the female but also the immediate vicinity of the nest.

Nest building

Nest site selection is done mainly by the female, although the male may accompany her in her search. Nests built over water among marsh vegetation are found in the primitive members of the Anatidae and the protection thus afforded from predators may have tipped evolution toward a more wholly aquatic existence. Many species nest on dry land, however, often, in the Anatini, several hundred yards from water. Some species, especially the shelducks (Tadornini), use crevices and holes for nest sites. They do not construct their own but take over what is available, such as rabbit burrows. Most hole nesters, however, utilize holes in trees, again preexisting ones, and will quite readily adopt nesting boxes where trees ancient enough to have holes are absent. Old tree nests of other species are sometimes used. A successful species like the mallard is adaptable in choice of nest site and may be found in any of the above situations.

The building of the nest is generally the work of the female, though the male assists in some species. It is a general rule that waterfowl are behaviourally incapable of carrying material in the bill, and even in the rare instance of courtship feeding shown by the red-crested pochard (*Netta rufina*), the female has to swim to the male to take his "gift." In most waterfowl, therefore, the nest is constructed of materials on the site within reach of the beak. They are pulled in and dropped sideways over the shoulder, forming a pile as the bird rotates. With the exception of species that share incubation, such as the magpie goose, the whistling ducks, and white-backed duck (*Thalassornis leuconotus*), the nest is progressively lined with down preened from the female's breast. This down is pulled over the eggs when the female leaves, keeping them warm and, in open sites, protected from the sight of predators.

Egg laying usually occurs early in the morning and occupies but a few minutes, though the female remains on the nest for up to an hour. Generally one egg per day is laid, with occasional gaps. The greatest spread recorded is in the velvet or white-winged scoter (*Melanitta fusca*), which takes 15 days to lay its usual clutch of nine eggs. Sperm has been shown to remain viable in a female mallard for a similar period, so it is probable that copulation prior to egg laying is sufficient to fertilize the whole clutch. Clutch size is variable within a species. The specific averages range from three to twelve eggs. There do not appear to be definite relations between clutch size and taxonomic, ecological, or geographical groupings, and its significance in anseriforms is still controversial. There is, however, a tendency for larger clutches to be laid by species that devote little parental care to their offspring. Egg size varies from 3 to 15 percent of the body weight in different species, the relatively largest being produced by the Oxyurini. In all cases the eggshells are immaculate white, green, or brownish. Surface finish ranges from very smooth to dull and large-pored.

The laying of eggs in nests other than their own is of widespread occurrence among waterfowl, and the very large clutches recorded, up to 22 in the mallard, are usually the work of more than one female. In the Aythyini and the Oxyurini such nest-parasitism is so common that it influences the breeding biology of a population. Only one species, the black-headed duck (*Heteronetta atricapilla*) of South America, is an obligate nest-parasite, always laying in the nests of other species.

With few exceptions, only the female incubates, the average period ranging from 22 to 39 days. Usually she leaves the nest periodically to feed, covering the eggs with down. In some species, like the common eider (*Somateria mollissima*), she apparently fasts during the four-week period. Her resulting constipation may be the cause of the extremely evil-smelling feces that she splashes over the clutch if suddenly flushed from it. This was once thought to deter predators, but experiments have not confirmed this. As incubation proceeds, aided by both the high temperature (around 101° F) and the high humidity maintained in the nest, the female becomes increasingly reluctant to leave the nest at the approach of a predator. If flushed late in incubation or with hatched young, she will indulge in injury-feigning, flapping or dragging her wings in an uncoordinated fashion that serves to distract predators.

Adaptations for short breeding seasons

The length of the incubation period shows little relation to taxonomic position or body size, but both it and the subsequent fledging period are correlated with the latitude at which breeding takes place. The continuous daylight of the Arctic summer enables the young to feed around the clock, but this does not sufficiently offset the fact that there is only a short interval between the melting of the snows and the onset of autumnal bad weather. Selection in the course of evolution has reduced the breeding sequence as far as possible. Thus the small high-Arctic races of the Canada goose (*Branta canadensis*) have so reduced the period from nest initiation to fledging that they can breed in areas having but two snow-free months. Larger races nesting in California take twice as long. Waterfowl are not double brooded in the wild, even when there is time enough. Most will renest, however, if the first clutch is destroyed. The second clutch is generally smaller but may have a higher fledging success in that hatching occurs in the less uncertain weather of late summer. Aviculturists regularly take advantage of this capacity, incubating the first clutch under bantam hens.

Since incubation does not start until the clutch is complete, hatching is generally well synchronized and may be complete within six hours. Another six hours is required for the down to dry, and young birds usually remain in the nest overnight. While in the nest they are brooded by the female and become imprinted with her visual and auditory characteristics. The downy young have sufficient internal food reserves to last two or three days if inclement weather prevents feeding as soon as they leave the nest.

The nesting behaviour of screamers is somewhat different from that of waterfowl. A substantial nest is built on marshy ground or in shallow water, sometimes reaching a height of 30 centimetres or more. The clutch varies from two to six eggs, depending on the species. Young screamers are covered with yellow down and leave the nest when a few days old, following their parents in the manner of young waterfowl.

Knowledge of waterfowl breeding, although far from complete, has come a long way from medieval times when barnacle geese (*Branta leucopsis*) were thought to hatch from barnacles on rocks because no one had then found their high-Arctic nests.

Ecology. Most waterfowl are associated with fresh water, at least during the breeding season. Relatively few inhabit the areas of early igneous rocks, which have deep lakes and fast streams of low productivity. More preferred are glaciated areas rich in shallow, productive lakes and marshes: the "pothole" country of North America, the silted deltas of rivers, the coastal marshes. An area with many small waters will support more ducks than a single large lake, because the extent of land-water margin is important. The longer shoreline produces diverse food organisms and provides isolation for breeding pairs, cover from predators, and shelter from winds. Paradoxically, periodic droughts are advantageous to duck production. The small lakes dry out and their bottoms become colonized with plants. When reflooding occurs, a rich diet of seeds and of invertebrate animals feeding on the decaying vegetation becomes available. Long-term or permanent droughts, such as occasioned by climatic changes or, increasingly, by man's drainage operations, can be disastrous. The widening acreage of artificial water reservoirs only partly offsets the loss, for they are normally large, deep waters, of little use to waterfowl except for roosting.

Migration

The ability to migrate greatly increases the ecological flexibility of waterfowl. They can exploit the summer resources of the northern tundra without having to evolve the encumbering adaptations needed to survive the winter there; they can move great distances to take advantage of local rainfalls in arid countries; they can retire to secluded areas when flightless and then seek the rich harvests of man-exposed environments when fully equipped to face the dangers. There are many patterns of migration, some regular and in fixed, reciprocal directions, others varying with changes in weather. The distances covered also vary widely. The longest waterfowl migrations are probably those of the blue-winged teal (*Anas discors*), which nests up to 60° N in North America and winters down to beyond 30° S, a distance of over 9,600 kilometres (6,000 miles). In the Old World the northern shoveler (*Anas clypeata*) has a similar migration distance of up to about 11,000 km. The northern pintail and the garganey (*Anas querquedula*), which breed in the Siberian tundra and taiga and winter in the tropical swamps of Senegal and Chad, are even more remarkable in their adaptability.

The general timing of regular long-distance migrations has been built into an internal hormonal cycle in the course of evolution. These "annual clocks" are kept in synchrony by changes in day length, but the immediate factors initiating the stages of migration are climatic. A mass southward movement occurs when air pressure systems produce a favourable wind flow; overcast conditions or a warm spell may temporarily halt the birds; bright, cold conditions move them on.

Causes of mortality

Sudden adverse weather may cause considerable mortality, especially where a long sea crossing is involved, as from Greenland to Britain. But the effects of weather on waterfowl populations can be most drastic at the time of breeding. Late frosts and spring floods are disastrous, especially when there is no time for renesting. Consequently, high-Arctic breeders may have near-complete reproductive failures. The brant geese (*Branta bernicla*) of the population wintering in western Europe only produce good crops of young in about half the seasons.

Parasites of many kinds inhabit waterfowl, generally in a state of tolerance, gaining the upper hand only when the bird is stressed in some way, as by food shortage. This is also true for fungal infections, particularly by *Aspergillus,* a common cause of death in captivity and, exceptionally, in the wild. Bacterial diseases take their toll and sometimes cause spectacular die-offs. Thousands of ducks have succumbed to botulism when shallow, brackish waters dry out in warm countries. *Clostridium botulinum* then multiplies, producing toxins that cause a fatal paralysis if ingested. Fowl cholera (*Pasteurella*) sometimes causes epidemics. More frequently nowadays mass mortality is occasioned by pollution, especially oil spills, pesticides, and other poisons. A peculiarly insidious danger is lead poisoning caused by the ingestion of lead pellets that accumulate in the mud of waters frequently shot over.

The direct kill by man undoubtedly makes him a major predator of waterfowl, and this must be increasingly controlled in the future. Adult ducks, other than nesting females, suffer relatively little natural predation, avian or mammalian, unless they are sick or weakened by hunger. The main predatory impact comes on the eggs and downy young. Crows and gulls, mink, raccoons, coyotes, foxes, ground squirrels, snakes, snapping turtles, bullfrogs, pike, and carp all take some toll. In normal circumstances such predation does not unbalance population dynamics, which are, after all, geared to a high rate of loss in the early stages. Predator-control measures have seldom produced any comparable increase in waterfowl production.

Beneficial relations exist with other birds and animals. Thus Aythyini and eiders (*Somateria* species) actually prefer to nest in association with the smaller gulls (*Larus* species). The latter tend to drive off predators and, by increasing the number of available eggs, share the pressure of egg predators. The muskrat clears water openings in reed beds and builds "houses" that are utilized by ducks for standing and nesting. Beaver dams create very attractive duck habitats. And, of course, man himself is not wholly destructive but creates some new areas for waterfowl and provides them with food, both unintentionally and on purpose.

FORM AND FUNCTION

Variety of bill modifications

Anatomy. The characteristic features of the family Anatidae are the skin-covered, lamellate bill and thick, fleshy tongue. The Anhimidae have a short non-lamellate bill, slightly hooked, like that of a domestic hen, and a horny tongue. Their bill is adapted for tearing aquatic plants. Among the Anatidae the basic bill has undergone a wide adaptive radiation. The geese have evolved strong, deep bills with hard, sharp lamellae. In some, such as the red-breasted goose (*Branta ruficollis*), the bill is short and slight, used only for grazing; in others, such as the snow goose (*Anser caerulescens*), it is long and heavy enough to deal with and dig for roots and tubers. The massive digging bill reaches maximum development in the magpie goose. The little pygmy geese (*Nettapus*) are so-called for their gooselike bills, but they actually feed on lotus seeds and water vegetation and do not graze or root for food. The European widgeon (*Anas penelope*), on the other hand, does graze extensively, but its bill differs little from the typical *Anas* bill, flattened with platelike lamellae. This is used for sifting particles out of mud or picking up food items from the lake bottom as the bird upends. The sieving bill is yet further developed in the shovelers and the pink-eared duck (*Malacorhynchus membranaceus*), the lamellae becoming extremely fine, enabling particles as small as diatoms to be taken from the surface film. The blue duck (*Hymenolaimus malacorhynchos*) has a rounded, expanded tip to the bill, which probably protects it when poking around sharp pebbles. The Aythyini have fewer lamellae and a narrower bill than the Anatini. In the mergansers (*Mergus* species) the lamellae have become toothlike projections in the long narrow bill, ideal for holding fish; additionally, the tongue has two dorsal rows of barbs. None of the Anatidae have developed pickax bills or ones for crushing hard food items, such as mollusks. When such are taken they are broken up in the gizzard.

Drawing by D. Malick

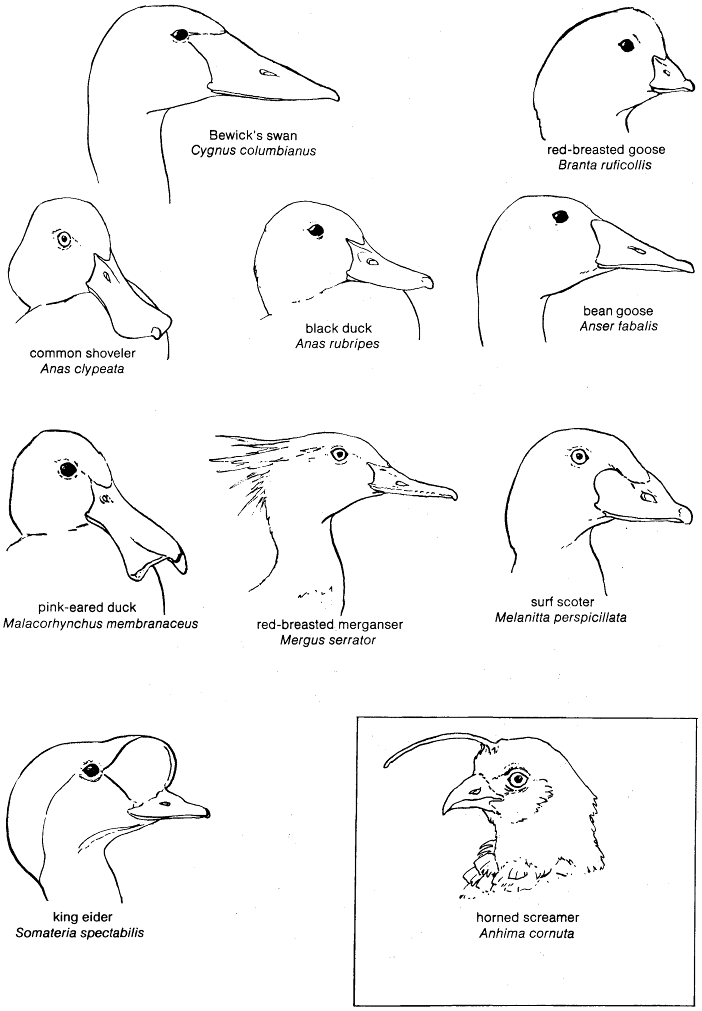


Figure 20: Heads of anseriform birds.

The legs and feet vary according to whether the main mode of progression is walking (geese) or swimming (divers). The latter have flattened tarsi to reduce water resistance on the forward stroke, the hind toe being lobed. Webs are fully developed between the front three digits, except in the magpie goose (primitively) and the Hawaiian goose, where it is believed that the greater flexibility of the toes makes for easier walking on broken lava beds. The claws on the digits are particularly sharp in forms (such as the Cairinini) that normally nest in tree holes.

The wings are not of unusual structure but vary widely in shape, from the broad expanse of the slow and manoeuvrable fliers to the narrow sweeps of the fast movers. Some of the Dendrocygnini have modified vanes on the outer primaries that produce a whirring noise in flight. The soughing creak of the wings of the mute swan (*Cygnus olor*) is well known. In a few cases the wing is so reduced in size that the bird is flightless, as is the Auckland Island race of the brown teal (*Anas aucklandica*). Many waterfowl have horny knobs on the carpal joint; in some there is a distinctly projecting spur as in the screamers, spur-winged goose, and torrent duck. In the first two, the spurs are actually used effectively in striking an adversary; in the last they may be of more use in progressing over slippery rocks in a raging stream.

Specializations of the windpipe

A structure that has received much attention, largely because of its taxonomic value, is the anatid trachea. This may be straight (geese and ducks) or looped in various ways (magpie geese and swans), the most elaborate being that of the trumpeter swan (*Cygnus cygnus buccinator*). Here it first enters the sternum, flexing twice into horizontal and vertical bony pockets, then emerging again to coil around and back into the lungs. The analogy to an orchestral wind instrument is obvious and a comparable sound is produced. Less obvious is the relation of shape to noise in the asymmetrical chamber (bulla) found at the base of the male trachea in most ducks. The male's whistle certainly differs from the female's quack, but it is difficult to account for the variation in bulla shapes—round and solidly ossified (Anatini), angular with membranous "windows" (Aythyini), or complexly membranous (Mergini).

The anseriform birds share with the curassows, tinamous, and ratites the possession of an intromittent male organ. This is not homologous with the mammalian penis but is a vascularized sac everted from the wall of the cloaca. It is protruded by muscular action and retracted by an elastic ligament. This gives it a spiral form, and sperm pass from cloacal papillae along a groove to the tip. This organ doubtless facilitates copulation on the water. Its presence is an aid to research and aviculture, since it permits the investigator readily to sex live birds that lack plumage dimorphism. The size of the organ also allows an estimate of the bird's age. That is, however, more safely done by determining the depth of the bursa of Fabricius, a glandular organ opening into the cloaca and diminishing with maturity.

Respiratory adaptations for diving

Physiology. There are interesting adaptations in the respiratory physiology of diving ducks. The accumulation of carbon dioxide does not stimulate breathing as it does in land mammals. When the bird is under water, oxygen is drawn from stored oxyhemoglobin and oxymyoglobin in blood and muscles, instead of from the respiratory system. The air sacs are compressed, reducing buoyancy. Glycogen is broken down by anaerobic and aerobic means, and the blood flow is modified so that the central nervous system receives the main supply.

Rather little is known about the physiology of digestion in waterfowl. It appears that, despite being almost entirely grazers, geese lack the ability to digest cellulose, either by the secretion of cellulases or by the symbiotic activities of microorganisms in the gut or its cecae (branches). Such inefficient digestion is probably correlated with the extraordinarily rapid passage of ingested material, which begins to appear in the droppings within half an hour and can be completely eliminated in two hours.

Salt excretion is achieved by the paired supraorbital nasal glands situated on indentations on the skull over the eyes. Races living in marine or brackish environments have these glands enlarged, quite modifying the profile of the head. In a botulism infection, it is probable that the toxins of *Clostridium botulinum* cause these glands to malfunction, resulting in the characteristic salt imbalance and paralysis.

EVOLUTION AND PALEONTOLOGY

Anseriform fragments have been recognized from as far back as the Eocene, and certainly by the late Oligocene birds assignable to the present-day tribes Anserini, Anatini, and (possibly) Oxyurini existed. The Miocene added Dendrocygnini and Aythyini, the Pliocene Tadornini, Mergini, and possibly Cairinini. By the Pliocene several species osteologically indistinguishable from modern species were present. Fossils from the Pleistocene are much more extensive and amount to more than 90 anseriforms. Paleornithologists tend to be conservative in allocating new specific names to Pleistocene anseriforms, since the observed differences correspond to those between modern races, indicating temporal as well as geographical continuity. Nevertheless, 32 of the Pleistocene species are reckoned to be extinct.

Despite the number of fossil species, the fossil record is inadequate to throw light on the origin and evolution of the Anseriformes. The greatest number and variety of their fossils have been reported from the Palaearctic Region, but this does not necessarily mean that the group originated there. Indeed, a number of other considerations point to a tropical origin, possibly in the Southern Hemisphere. At present, the greatest number of endemic and monotypic genera and the more generalized species are to be found there. Be that as it may, the Anseriformes soon spread and were to be found over all their present range quite early in their history.

Main lines of anseriform evolution

Ancestral forms were probably long-legged unspecialized birds with features related to the galliform curassows (Cracidae), ciconiiform flamingos (Phoenicopteridae), and anseriform screamers and magpie goose. While the screamers have probably advanced but little, the Anatidae have diversified rapidly and extensively. Two main lines may be distinguished. One, typified by the swans and geese, became large-bodied, long-lived, slow-breeding birds, with prolonged pair and family bonds and associated parental care of the young. The other line, characterized by most of the ducks, has favoured small-bodied, short-lived birds having but brief pair bonds and minimal care of the young, necessarily correlated with prolific production of young. Superimposed on these general trends were adaptive radiations; from omnivorous to specialized food habits; from a largely terrestrial to an almost fully aquatic existence. An adaptation for the latter, the double molt, enabled the males to develop striking nuptial plumage, which prevented interspecific hybridization when many species were sympatric. Migratory habits strengthened this requirement, and gene flow within such species was rapid. Racial differences developed only in nonmigratory or otherwise isolated species, or in those migratory forms that maintained family groupings when away from the breeding grounds, and so developed learned, traditional attachments to the area where they were born.

Despite the wide variety of present-day Anatidae, the family is basically genetically homogeneous. This is shown by the readiness with which hybrids occur, at least in the forcing conditions of captivity. More than 400 kinds of hybrids are known. Many of the interspecific crosses are fertile, and some remarkable intergeneric and intertribal hybrids are known. The evidence from serological investigations and from the physico-chemical examination of egg-white proteins and egg-shell structure similarly shows the family to be close-knit, only the magpie goose deserving classificatory isolation. Likewise the eggshell structure and nitrogen content throw some light on evolutionary relationships. Feather pigments pose a number of fascinating problems, and recently the composition of feathers has been analyzed to throw light on the geographical origins of goose populations.

CLASSIFICATION

Distinguishing taxonomic features. Broad classification divisions are based on skeletal and other internal structures. Finer divisions are decided on plumage type and patterns, including those of the downy young, and on behavioural features such as the length of pair bond and on displays associated with pair formation and maintenance.

Annotated classification. The classification here follows that of Jean Delacour (1954) as modified by Paul Johnsgard (1965, 1968).

ORDER ANSERIFORMES

Birds with skull desmognathous (maxillopalatines fused) and holorhinal (posterior border of nostrils rounded); one pair of posterior sternal notches; four digits on each foot; wing diastataxic (gap between secondaries four and five); 11 primaries, the outer one reduced; oil gland tufted; eggs unmarked; young nidifugous, with thick down.

Family Anhimidae (screamers)

Pleistocene-Recent. Short, curved, nonlamellate bill; ribs without uncinate process; feet unwebbed, hallux long; plumage uniformly distributed; extensive air cells between skin and body. South America; length 70–95 centimetres, weight 2–5 kilograms; two genera, three species.

Family Anatidae (ducks, geese and swans)

Eocene-Recent. Skin-covered bill, lamellate edges, tipped with horny nail; occipital fontanelles generally present; hallux generally reduced and raised; tongue thick, bordered with spiny processes; primary wing feathers generally molted simultaneously; male copulatory organ present; body length about 30–150 centimetres, weight 300 grams–18 kilograms; worldwide; about 37 genera, 142 species.

†*Subfamily Eonessinae.* Fossil. Known only from middle Eocene of Utah, one species.

†*Subfamily Romainvilliinae.* Fossil. Known only from lower Oligocene of France. One species.

Subfamily Anseranatinae (magpie goose). Rear of head slopes upward and forward; metasternum three-lobed; circular foramen in precoracoid bone; short descending process on lachrymal; anterior toes half webbed, long hallux not raised; long trachea coiled beneath skin; gradual molt of primaries; feeds young from bill; shares features with Anhimidae and with Anserinae. Northern Australia; one genus, one species.

Subfamily Anserinae (swans, typical geese, and whistling ducks). Legs short to medium; neck medium to long. Anterior toes webbed, hallux short, elevated, tarsal scales reticulate. Sexes much alike in plumage, voice, and behaviour. Long pair bond and extended family life. Single annual molt. Vegetarian. Wings usually moderately long, pointed; tail short. All species swim capably, many primarily aquatic.

Tribe Dendrocygnini (whistling ducks). Whistling voice. Relatively long-legged, with upright stance. Enclosed bony orbit (eye socket). Strongly patterned adult plumage; downy young with light stripe below eyes and around nape. Mature at one year. Mutual postcopulatory step-dance. Widely distributed in tropics and subtropics. Two genera, nine species.

Tribe Anserini (swans and geese). Simple trachea and syrinx. Mature in second to fourth year. Behaviour includes "triumph" ceremony and precopulatory head dipping. Breeding distribution Holarctic and north temperate, except for one tropical goose and two species of south temperate swans. Five genera, 21 species.

Tribe Stictonettini (freckled duck). Ducklike with short legs but many anserine features. Australia. One monospecific genus.

Subfamily Anatinae (ducks and aberrant geese). Tarsal scales scutellate (overlapping plates). Male trachea with asymmetrical enlargement (bulla); sexes differ in voice, and also usually in plumage and behaviour; brief pair bond and family life. Double annual molt; males often alternate a cryptic "eclipse" plumage with a conspicuous nuptial one; downy young strongly patterned. Female "inciting" display present.

Tribe Tadornini (sheldgeese and shelducks). Iridescent green wing speculum and, generally, white wing coverts; sexes alike in majority; mature in second year. Pugnacious. Downy young black and white. Worldwide except North and Central America. Five genera, 15 species (one extinct).

Tribe Tachyerini (steamer ducks). Ponderous birds with short wings; plumage mottled with no metallic colours, sexes alike. Closely related to Tadornini, but differ in display behaviour. Western and southern South America. One genus, three species.

Tribe Cairinini (perching ducks). Heterogeneous group; perching habit; metallic wing plumage; small species sexually dimorphic in plumage, larger ones in size; hole nesters; downy young sharp-clawed and stiff-tailed; pair bond usually short-lived or barely developed. Primarily tropical and subtropical; some temperate. Nine genera, 13 species.

Tribe Anatini (dabbling ducks). Short-legged; lobed hallux. Tracheal bulla round and ossified; mature in first year; iridescent wing specula; female plumage with disruptive patterning; female inciting behaviour very ritualized; female has distinctive decrescendo call; mutual head pumping as precopulatory display; complicated group displays in males. Downy young with eye and cheek stripes. Most feed on water surface or by upending. Worldwide. Five genera, 40 species.

Tribe Aythyini (pochards and scaup). Freshwater divers, mainly vegetable feeders, with large feet set well back, and strongly lobed hallux. Tracheal bulla angular with membranous "windows"; sexual dimorphism in plumage, but male nuptial plumage not strikingly patterned; mature at one year. Females without decrescendo call; courtship behaviour simple; downy young lack facial stripe. Worldwide distribution. Three genera, 16 species (one extinct).

Tribe Mergini (sea ducks and mergansers or sawbills). Mostly saltwater divers, feeding on animals. Bulla structures often elaborate; sexual dimorphism in plumage, voice and behaviour; male courtship patterns very varied; maturity requires at least two years. Typically Holarctic, one species in South America. Eight genera, 20 species (two extinct).

Tribe Oxyurini (stiff-tailed ducks). Freshwater divers with legs so far to rear that walking is difficult; lack tracheal bullae, but sound production by inflatable sacs in neck; sexual dimorphism in dense, shiny plumage, male patterns not elaborate; long, stiff tail feathers; elaborate male displays; large eggs and precocial young. Mainly Southern Hemisphere. Three genera, eight species.

†*Subfamily Paranyrocinae.* Fossil. Known only from lower Miocene of South Dakota. One species.

Critical appraisal. There is general agreement on the family and subfamily divisions of the order and at the generic level, where most workers agree on the lumping of species into large genera. Some controversy exists at tribal level. Some workers would include the Tachyerini within the Tadornini; extract the eider ducks from the Mergini and erect for them the tribe of Somaterini; reorganize the admitted "ragbag" of the Cairinini; eliminate Stictonettini and return the freckled duck (*Stictonetta naevosa*) to the Anatini. Indeed it is on the positioning of certain "difficult" species that most discussion revolves. Certain species whose taxonomic positions are in dispute are best thought of as linking species, the result of evolutionary lines that diverged early from near the bases of two major phyletic lines. Examples of such species are the white-backed duck (*Thalassornis leuconotus*), which seems intermediate between the Oxyurini and Dendrocygnini; the Cape Barren goose (*Cereopsis novaehollandiae*), placed with equal justification in the Tadornini or the Anserini; the ringed teal (*Callonetta leucophrys*), sometimes placed in the genus *Anas*, in the Anatini and sometimes the Cairinini; the pink-headed duck (*Rhodonessa caryophyllacea,* extinct since about 1940), placed either in the Anatini or in the Aythyini; the coscoroba swan (*Coscoroba coscoroba*), assigned to the Dendrocygnini or the Anserini; and the marbled duck (*Marmaronetta angustirostris,* sometimes placed in *Anas*), which is evidently intermediate between the Anatini and Aythyini. The study, in recent years, of new taxonomic characters such as behaviour patterns and biochemistry has substantially improved our understanding of relationships within the Anseriformes.

(G.V.T.M.)

Ciconiiformes (herons, storks, ibis, spoonbills, flamingos)

The Ciconiiformes constitute an order of five or six families of storklike birds: herons and bitterns (Ardeidae), the shoebill (sole species of the Balaenicipitidae), the hammerhead (sole species of the Scopidae), typical storks and wood storks (Ciconiidae), ibis and spoonbills (Threskiornithidae), and, according to some authorities, flamingos (Phoenicopteridae). Most are of substantial size, long-legged and long-necked, and adapted for wading. They are widely distributed, often abundant, and apt to be conspicuous in their open habitats or in the air. Many are notably graceful in form and movement, and some have spectacular powers of flight.

GENERAL FEATURES

Size range and diversity of structure. Some storks are very large, standing over 1.2 metres (four feet) high and having wingspans of up to 2.6 metres (8.5 feet). The larger herons are about as tall when standing erect. Flamingos are also tall, with great length of neck and legs. Species of medium size usually stand 60–90 centimetres (2–3 feet) high and some of the smaller ones as little as 30 centimetres (12 inches). Exceptionally small are the little bitterns of the widespread genus *Ixobrychus,* weighing less than 100 grams (a few ounces).

With the partial exception of the flamingos, the structural characteristics of the order are well-marked, and the same is true of the families. The storks, even including the wood storks, form a recognizable group of birds of from medium to large size. The ibis form an even more homogeneous group, birds of medium size with markedly down-curved, slender bills; the spoonbills, of the same family, differ in this respect, as the name implies. The herons are more diverse, with a greater size range, the bitterns standing a little apart in behaviour more than in structure. Each of the remaining two families contains a single species with some peculiar characteristics.

Distribution, habitat, and abundance. Ciconiiforms are found throughout the world, except in the polar regions, but the largest number of species is found in the warmer parts. Some of those breeding in the North Temperate Zone perform long migrations. The herons are the most cosmopolitan family, some being found even on remote oceanic islands. The shoebill (*Balaeniceps rex*) has a limited distribution in tropical Africa, and the hammerhead (*Scopus umbretta*) is almost confined to that continent. The typical storks (subfamily Ciconiinae) are not represented in North America and have only one species in the Australasian Region; the wood storks (subfamily Mycteriinae) are represented by one species in the New World and three in the Old World. The ibis are widely distributed, but are not found in New Zealand. The flamingos are found throughout the tropics, but are not represented in the Australasian Region.

The usual habitat of ciconiiforms is near water, chiefly fresh, and only a few species, such as the white stork (*Ciconia ciconia*), live largely on dry ground. The flamingos require brackish or alkaline water, and two species inhabit Andean lakes at elevations of up to about 4,000 metres (14,000 feet).

Threatened species

Many of the species in the order are very abundant, with world populations running into millions, and some assemblages, such as those of the lesser flamingo (*Phoeniconaias minor*) in Africa, are enormous. At the other extreme, the Japanese ibis (*Nipponia nippon*) is on the verge of extinction, only one small colony being known. Several other ibis species are rare and are declining in population.

Relations with man. Although some freshwater fishing interests may regard herons as undesirable competitors, on the whole the members of this order are considered to be either beneficial or neutral in respect to the human economy. They are not sought by man, and many are protected. At one time a number of species of little egrets were persecuted at their breeding places for the sake of their nuptial plumes—aigrettes (or ospreys)—but such killing has now largely been stopped.

The association of the white stork (now an endangered species) with mankind in its preference for nesting on buildings is further mentioned below, but it is not the only species thus involved. Abdim's stork (*Sphenorhynchus abdimii*), for instance, will nest on native huts in a treeless area. And many of the arboreal colonies of heron and stork species in Africa are in or near villages. The cattle egret's dependence on domestic stock to flush insects, as an alternative to wild herbivores, brings it into a familiar relationship with human beings.

NATURAL HISTORY

Feeding habits. Most ciconiiforms subsist wholly or mainly on animal food, which is usually swallowed whole, indigestible matter being regurgitated later as pellets. Fish are the prey of many species such as the larger herons, but small mammals, birds, reptiles, and amphibians are also taken, as well as invertebrates, including mollusks, crustaceans, insects, and worms.

The flamingos, on the other hand, live chiefly on minute algae.

Specialized feeding of herons

The methods of fishing are varied. Some of the larger herons stand solitary and motionless until a fish comes within reach and the long neck suddenly shoots out. Certain herons employ a hunting method known as "canopy feeding," in which one or both wings are held forward, forming a canopy over the head and creating a patch of shaded water. It is thought that fish mistake the shaded area for a safe refuge; another interpretation is that the canopy aids the bird's vision by reducing surface reflections. Other herons go actively in pursuit, stirring up the bottom with their feet or wading rapidly about. The night herons are largely nocturnal in their fishing. Some members of the heron family are not fish eaters. The cattle egret (*Ardeola ibis*) lives on insects caught when disturbed by the large mammals with which the bird associates.

The food of storks is also varied and may be sought on dry land, in marshy ground, or in shallow water. White storks wintering in Africa take large numbers of locusts when the latter are swarming. Wood storks fish by plowing the water, with mandibles open and partly submerged, as they wade in the shallows. The marabou (*Leptoptilos crumeniferus*) and other members of the same genus are partly predatory but largely scavengers and carrion eaters, and they may often be seen with vultures at a carcass.

Ibis use their long bills for probing the ground or mud. The spoonbills catch small organisms by moving their bills, with mandibles slightly open, from side to side in shallow water. The hammerhead often hunts for frogs and will use the back of an almost submerged hippopotamus as a vantage point. The shoebill is a fish eater.

Filter feeding in flamingos

Flamingos have a highly specialized method of feeding, which has its closest parallel in certain whales. The inside of the mandibles carries a series of fine hairlike plates that act as filters when water is sucked into the mouth and then expelled. The head is held with the forehead downward. The food particles caught on these plates are ultimately worked on to the tongue and from there are swallowed. There are differences of detail between the species.

Vocalization. Many of the ciconiiforms are rather silent, making at most various croaking or grunting sounds. Some storks utter scarcely any sounds, but often noisily clatter their mandibles. Some herons have loud, harsh cries. Among the ibis, the hadada (*Hagedashia hagedash*) of Africa is exceptional in being noisy—flying about with yelping cries.

In spring, the males of the large bitterns make an extraordinary booming sound audible at up to five kilometres; for this purpose, the esophagus is modified so that it can be inflated and serve as a resonating chamber.

Migration. The relatively few species breeding in northern areas with hard winters are summer visitors there. In some herons there is a tendency to a northerly postbreeding dispersal before the time of true migration southward.

The white stork of Europe and northern Asia is a typical long-distance migrant, travelling as far as South Africa and India. The black stork (*Ciconia nigra*) performs a similar migration, but it is doubtful that visitors to Africa now penetrate beyond the Zambezi River in Mozambique. In South Africa, on the other hand, there is a sparse resident population believed to have been derived, within the twentieth century, from migrants that formerly made a longer journey. There are migrants even among species confined to the tropics; Abdim's stork breeds in the northern tropics of Africa during the rains and migrates across the Equator.

Apart from regular movements, recent years have seen a notable range expansion by the cattle egret from Africa and Asia to the Americas and to Australia.

Behaviour. An outstanding feature in ciconiiform behaviour is gregariousness. Even where the mode of obtaining food necessitates solitude and therefore dispersal, the tendency is for reassembly at the end of the day and then flight in formation to a communal roost. Breeding is mostly colonial. Species able to feed in flocks are gregarious at all times. Some of the assemblies are enormous; for example, over 1,000,000 lesser flamingos occur on a single African lake. There may be a seasonal element, as when white storks, dispersed in the breeding season, form flocks for migration.

There are some exceptions. Bitterns and tiger herons, relying on concealment for protection, are not markedly gregarious, nor is the hammerhead. The shoebill is usually seen singly or in pairs. It is rather silent, partly nocturnal, and has even been credited with a morose disposition.

Courtship behaviour of herons

Behaviour related to mating and the pair bond is well marked and has been described in detail for several species, notably of herons. Among these, the first signs of breeding may be the assembly of birds on a gathering ground adjacent to the colony and the performance there of various strutting and dancing rituals. A male may then take off on a "circle flight" ending at the site that he has chosen for the future nest. There he may adopt a series of stereotyped display postures, such as the "stretch," in which he first extends the head and neck vertically and after a moment bends them rearward until the head is almost touching the back. There may also be a "snap" display, in which the head and neck are extended forward and downward, the feathers of the neck, face, and crest are erected, and the mandibles are brought together with a loud clap. After the male heron has attracted a mate, various mutual displays during building, or at nest relief during incubation, apparently help to maintain the pair bond. Both partners will defend their territory against encroachment by others of their kind, this defense involving various types of threat display.

Analogous sexual and hostile display rituals occur in other families of the order, but in the main have been less intensively studied than in certain herons. In a well-known greeting ceremony used by various storks, when mates are reunited at the nest after one has been absent, the neck is bent back so that the head rests upside down on the body with the bill pointing toward the tail. Occasional mass aerial manoeuvres above the nesting colony are performed by wood storks and others.

Locomotion. On land or in shallow water these birds commonly walk with deliberate gait, but with rapid strides when necessary. All, except the flamingos, ordinarily perch on trees or, in a few cases, among reeds, most species roosting and nesting there. Bitterns and small herons often climb with agility among reeds or branches. Only the flamingos commonly swim when the water is too deep for wading, although some others immerse themselves to some extent while catching food.

The flight of ciconiiforms is strong, on broad wings, with the legs trailing behind. The neck is drawn back in herons, slightly so in the hammerhead; in the shoebill it is fully drawn in with the heavy bill resting on the breast, as in pelicans. In the other families, with the exception of storks of the genus *Leptoptilos*, the neck is stretched forward. In direct flight, flocks often assume V-formation. The larger storks perform impressive feats of soaring in thermal upcurrents.

Reproduction and nesting. The breeding of ciconiiforms tends to be in large, sometimes enormous, colonies, which may comprise a mixture of species. Exceptions are tiger herons, bitterns, the shoebill, and the hammerhead, birds not markedly gregarious at any season.

Considering the feeding environment essential for most

Common egrets, *Egretta alba* (top left and right), with snowy egrets, *Egretta thula* (centre), in a typical colonial nesting situation.
Allan D. Cruickshank—The National Audubon Society Collection/Photo Researchers

ciconiiforms, it is remarkable that so many of them should be arboreal nesters. The commonest site is in a tree and often at a great height. Exceptions include the bitterns, which nest among reeds; the shoebill, which nests on the ground; and the flamingos, which also nest at ground level, often in shallow water. Among the storks the choice is more varied. Most species nest in tall trees, but some use cliffs or buildings. In parts of Europe, as an alternative to buildings, the white stork often uses artificial platforms erected for its benefit, commonly a cartwheel lying horizontally on top of a high pole, or a similar structure. The site is variable in night herons, ibis, and spoonbills, whose nests may be in trees, in low bushes, among reeds, or even on the ground (ibis and spoonbills), or on cliffs (some species of night herons).

Nest structure

The nest itself is usually a loose platform constructed of sticks and stems. Rushes are frequently used in marshy situations. The nest of the hammerhead, placed in a tree and often at no great height, is an enormous structure of sticks and other vegetation, forming a closed chamber sometimes a metre or more in diameter with an entrance tunnel at one side. The cavity is lined with mud or dung. The shoebill, in contrast, does no more than flatten a patch of long grass on dry ground. Flamingos build cones of mud on the lakeshore, 15–35 centimetres (6–14 inches) high, with a shallow depression scooped out of the top of each. The incubating bird sits on top of this nest, with its long legs folded beneath its body. Sometimes, as on small rocky islands, other materials are used in the absence of mud.

As a rule, flamingos lay only a single egg and the shoebill one or two. The usual clutch is three or four in ibis and spoonbills, three to six in storks and the hammerhead, three to seven in herons, and four to six in bitterns. There is a tendency for the eggs to be ovoid (*i.e.,* with the ends equally rounded) and to have a white chalky outer layer, underneath which there may be a coloured shell. The colours range from white to pale blue, green, or buff in most groups, to olive brown in the bitterns, and to dark greenish-blue in some ibis. The ground colour of ibis eggs may be spotted or blotched with brown, but the eggs of most others are plain. There are faint or scanty markings in a few.

The chicks are downy when hatched, or quickly become so, but, except for those of flamingos, they remain in the nest, dependent on their parents for food until full-grown. In the herons there is only one down plumage, but in storks and ibis there are two in succession. The downy young of flamingos are gray, with bills and legs that are red at first, but become black within a few days. After the first two or three days, the flamingo chicks leave the nest and move freely about the adjacent part of the colony. As they grow older they are herded in groups by a few adults, but each apparently continues to be fed by its own parents. The bill acquires its specialized form only as the young flamingo grows. Among some other members of the order, there is a tendency for the bill to be shorter and of a more generalized type during the early stage of postnatal development.

So far as information is available for different species, ciconiiform birds first breed when from two to four years old. Parental care is usually shared between the sexes, both taking part in building the nest, incubating the eggs, and feeding the chicks. Bitterns of the genus *Botaurus* seem to be exceptional in that only the female incubates the eggs and tends the young. Feeding is usually by regurgitation. In herons, for example, the parent's bill is seized crosswise by the young bird, and the food is passed sideways to its mouth from between the adult's mandibles. In flamingos, the regurgitated food is of fluid consistency, and the young must continue to be fed in this way until their own filtering apparatus is adequately developed at the age of some two months.

Feeding of nestlings

FORM AND FUNCTION

General characteristics. The carriage of the body is markedly upright in many species, but more horizontal in others. Length of neck and length of legs tend to be correlated; but the chestnut-bellied heron (*Agamia agami*) of tropical America combines short legs with a long neck. In herons, particularly, the neck is curled back on itself at rest but can be instantly straightened on alert or to seize prey.

Long legs and toes are often an adaptation for wading. The toes of herons are long and flexible, with just a hint of webbing for walking or standing on soft ground, the functional hind toe aiding in perching. The hind toe of storks and ibis is reduced and elevated, an adaptation for more walking and less perching. The relatively short front toes of the flamingos are completely webbed and the hind toe small or absent; these birds walk extensively on soft bottom mud. In the herons, and also in the hammerhead, the claw on the middle toes is serrated (pectinate) on its inner border, an aid to preening.

Bill shape. The bill shows various adaptations to methods of feeding (Figure 21). In herons it is typically long, straight, and pointed; although spearlike in appearance, it is used for grasping rather than impaling. In the boat-billed heron (*Cochlearius cochlearius*) the bill is flattened into a broad scoop and is used as such in feeding. In bitterns, both mandibles are finely serrated near the tip. The bill of the hammerhead is relatively short and slightly hooked. The shoebill has a large head with a heavy, bulging, hooked bill; this may be an adaptation to digging lungfish out of the mud but is used also for catching other fish.

In storks, the bill is usually long and strong, as in herons, but not always straight; it may be decurved as in the wood storks, or slightly upturned as in the jabiru (*Jabiru mycteria*) of tropical America. In the two openbills of the genus *Anastomus,* when the mandibles are closed a wide gap remains between their outer portions, except at the tips, probably in adaptation to holding large water snails. The ibis have long, thin, and markedly decurved bills; and in the spoonbills the tip of the long straight bill is flattened and broadened, thus forming a thin spatula.

In the flamingos the bill is bent downward in the middle of its length and has the filtering mechanism already described.

Plumage and coloration. Among the herons, plumages in patterns of blues, grays, and white are common; some, such as various egrets, are all white. In storks, contrasting areas of black and white are characteristic. Some of the ibis are notable for the metallic sheen on the feathers. The hammerhead is of a sombre brown hue. The large bitterns and the tiger herons have cryptic coloration, which is associated with their habit of standing motionless among the reeds with their bills pointing toward the sky.

Ornamental coloration

Touches of bright red on the plumage or on bare parts are not infrequent. A few species have brightly coloured plumage, such as the scarlet ibis (*Eudocimus ruber*) and the roseate spoonbill (*Ajaia ajaja*), both of tropical America, and, in varying degree, the flamingos.

The sexes are usually alike in plumage, or nearly so, exceptions being the little bitterns. The immature may be like the adult or quite different; in the scarlet ibis it is dull brown. Some herons have, in both sexes, long plumes

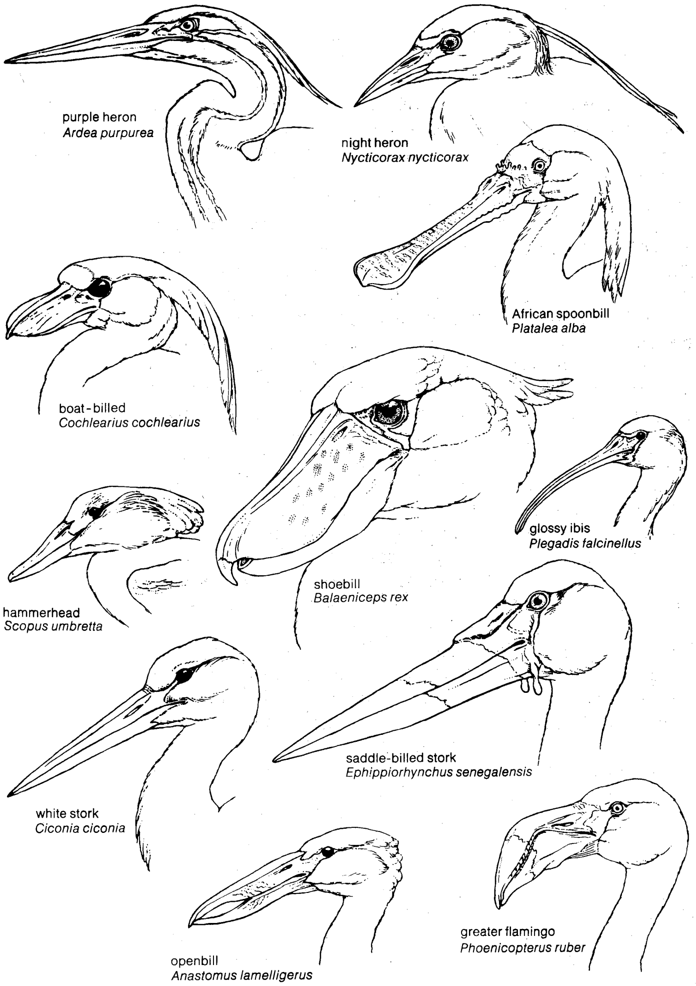


Figure 21: Variation in beak structure among ciconiiform birds.
Drawing by R. Keane

descending from the head, neck, breast, or back, in most only during the breeding season. Some species (*e.g.,* certain ibis, the boat-billed heron, and many typical herons) have crests on the crown or nape. The hammerhead gets its name because the bill appears balanced by an erectile tuft of feathers projecting backward. Patches (two or more pairs) of powder-down feathers are especially characteristic of the herons. These feathers break down to produce a fine powder, which is distributed to the plumage with the bill in preening.

The head and neck are partly or wholly bare of feathers in some species, and the exposed skin may be black, as in the sacred ibis (*Threskiornis aethiopica*), or brightly coloured red in the hermit ibis (*Geronticus eremita*) and others. In the marabou, the head and neck are almost naked and from the front of the neck extends a distensible pouch of pink skin.

Seasonal change in plumage is seen mainly in the shedding of the ornamental breeding plumes of some herons, but there are some changes in the colour of the bill, irises of the eyes, legs, and bare patches of skin during the season of sexual display. At this time the colours become brighter, with a common tendency toward red, and there may even be an intensification of the hue in moments of excitement.

Another characteristic of the herons, notably the egrets, is a tendency to dimorphism, resulting in a population containing both white and slate blue individuals.

EVOLUTION AND PALEONTOLOGY

The Ciconiiformes are generally accepted as being a primitive group. Apart from the equivocal position of the Phoenicopteri, and with possible reservations about *Balaeniceps,* there seems to be little doubt that the order evolved through a single phyletic line. The two main branches of the Ciconiiformes produced both the herons and the storks, with the remaining forms probably nearer the latter.

In the fossil record, the first heron, *Proherodius,* dates from the lower Eocene of England (about 54,000,000 years ago), and other herons from later in the Eocene of North America. The first ibis (*Ibidopsis*) is from the upper Eocene of England and the first stork from the late Eocene or very early Oligocene of France. From the Pliocene of North America (about 7,000,000 to 2,500,000 years ago) there are fossil forms of modern heron genera (*Ardea* and *Nycticorax*) and a modern ibis genus (*Eudocimus*).

The flamingos and their extinct allies have a long record, beginning with *Scaniornis* from the lower Paleocene (about 65,000,000 years ago) in Sweden. The first true flamingo (*Elornis*) is from the late Eocene of England, and another flamingo-like form, *Agnopterus,* from the same level; *Paloelodus* dates from the lower Miocene in France (about 26,000,000 years ago). Fossil families have been named for each of these three extinct genera.

CLASSIFICATION

Distinguishing taxonomic features. The members of the order possess in common, but not exclusively, a number of anatomical characters that are considered to be conservative (*i.e.,* changing only slowly in evolution), indicative of kinship in descent, rather than convergent evolution. These features include the arrangement of the palatine bones, the presence of diastataxy (*i.e.,* the wing apparently lacks a secondary feather associated with the fifth secondary covert, which is present), the presence of intestinal ceca (blind pouches), nearly always minute, the possession of 16–20 cervical (neck) vertebrae, and the presence of only a single pair of sternotracheal muscles in the syrinx (voicebox).

Annotated classification. The following taxonomic arrangement is widely accepted by students of this order. Groups marked with a dagger (†) are known only from fossils.

ORDER CICONIIFORMES

Suborder Ardeae

Family Ardeidae (herons, egrets, and bitterns)

Lower Eocene to Recent. Loose-plumaged wading birds of moderate to large size, most with slim body and long neck; bill usually long, straight, and sharp; legs medium to long, lower tibiae bare; partial web between outer and middle (sometimes also between middle and inner) toes, and hind toes at same level as front toes; claw on middle toe pectinate (*i.e.,* with comblike inner edge). Many with ornamental plumes in breeding season. Adult length 28–142 cm (11–56 in.). About 75 species (and 17 fossil species) usually divided into 2 subfamilies: Ardeinae (typical herons, egrets, tiger herons, night herons) and Botaurinae (bitterns). Worldwide, except in polar regions.

Suborder Balaenicipites

Family Balaenicipitidae (shoebill or whale-headed stork)

A large wading bird, stoutly built, with moderately long neck and short tail. Bill large, broad, and flattened, with hooked tip on upper mandible. Legs long, hind toe on same level as others, claw on middle toe slightly pectinate. Plumage sombre; short bushy crest. Length 117 cm (46 in.). One species, confined to tropical east central Africa.

Suborder Ciconiae

Family Scopidae (hammerhead, hamerkop, or hammerheaded stork)

A moderate-size bird with large head, short neck, rather long wings, and moderate length tail. Bill medium length and laterally compressed, straight, and slightly hooked at the tip. Legs long; toes slender, with partial web connecting front three; hind toe at same level. Plumage brown, substantial crest projecting horizontally from back of head. Length 50 cm (20 in.). Africa, southwestern Asia, Madagascar.

Family Ciconiidae (storks)

Upper Eocene to present. Medium to large wading or walking birds, stoutly built, with long necks. Bill long and massive, straight or curved up or down (or with gap between closed mandibles); legs long, with partial web between middle and outer toes, hind toe smaller and raised. Plumage often boldly patterned in black and white; parts or whole of head bare in some species. Length 76–152 cm (30–60 in.); 21 fossil and 17 Recent species, divided into 2 subfamilies. Ciconiinae (typical storks) and Mycteriinae (wood storks). Worldwide in warm regions, but most species in the Old World.

Family Threskiornithidae (ibis and spoonbills)

Upper Eocene to present. Medium to large wading or walking birds, with long neck and short tail; bill long, slender, curved downward (ibis), or straight and spatulate at the tip (spoonbills). Legs long; front toes slightly webbed at base, hind toe small and elevated. Many with crests; some ibis with whole or part of head and neck bare. Length 48–107 cm (19–42 in.). Seven fossil and about 28 Recent species, in 2 subfamilies. Threskiornithinae (ibis) and Plataleinae (spoonbills). Virtually worldwide in tropics and subtropics, a few species in temperate regions.

†*Family Plegadornithidae*

Fossil only; 1 species; Upper Cretaceous of Alabama.

Suborder Phoenicopteri

†*Family Torotigidae*

Fossil only; 3 monotypic genera; Cretaceous of western Europe and western North America.

†*Family Scaniornithidae*

Fossil only; 1 species; lower Paleocene of Sweden.

†*Family Telmabatidae*

Fossil only; 1 species; lower Eocene of Argentina.

†*Family Agnopteridae*

Fossil only; 3 species; upper Eocene of England and France, and upper Oligocene of Kazakhstan, U.S.S.R.

Family Phoenicopteridae (flamingos)

Upper Eocene to present. Very tall wading (and swimming) birds, with slender bodies, long thin necks, large wings, and short tails. Bill stout, bent sharply down at midpoint, and furnished with lamellar filtering apparatus. Legs very long; front toes relatively short and fully webbed, hind toe small or absent. Plumage mainly white, tinged pink, or light vermilion; face bare. Length 91–122 cm (36–48 in.). Nine fossil and 4 Recent species; shallow lakes and lagoons in tropics; worldwide, except Australasia.

†*Family Palaelodidae*

Fossil only; 8 species, lower Miocene to lower Pliocene; France, Germany, and western United States.

Critical appraisal. There is little controversy about the limits of the order Ciconiiformes apart from the question of whether the flamingos (Phoenicopteri) should be given separate ordinal rank (Phoenicopteriformes). Flamingos have affinities with Anseriformes and Ciconiiformes. There is not much dissagreement about the division into families, except that some place the boat-billed heron (*Cochlearius*) in its own family, Cochleariidae. Apart from the shape of the bill and associated skull modifications, clearly adaptive features, the characters seem to be mainly of the Ardeidae. Superfamilies have been erected from the three families of the Ciconiae; but with only one family in each, the device merely emphasizes their distinctiveness. On the other hand, some do not place the Balaenicipitidae in a separate suborder, but with the Ciconiidae as the superfamily Ciconioidea.

The precise position of the two species in monotypic families has been debated. The strongest affinities of *Scopus* appear to be with the suborder Ciconiae, in which its family is here placed, but it also shows resemblances to the Ardeae and the Balaenicipites. *Balaeniceps* was at one time placed in the order Pelecaniformes, and further osteological evidence tending in this direction has recently been adduced; others attribute the resemblances to convergence. Within the Ciconiiformes, the shoebill has affinities with both the Ciconiae and the Ardeae. (A.L.T.)

Falconiformes (falcons, hawks, vultures, and allies)

More than any other group of birds, the diurnal birds of prey, order Falconiformes, have long interested men. They are often large, conspicuous birds of swift, graceful flight, attracting attention from the smaller birds. Falconiforms compete with man for food by occasionally preying on his domestic stock (though on a small scale). But more significantly, they feed upon pests of human food supplies. Because of their predatory, or raptorial, skill, a few falcons and hawks are used for falconry, mainly for sport, much less importantly to obtain food.

GENERAL FEATURES

Habits and appearance

Falconiforms are seldom abundant but may be common or widespread, occurring in varied habitats. All are diurnal, none truly nocturnal, though several are active at dawn and dusk. With few exceptions they feed on the flesh of other animals, taken alive or found dead. Although some, especially in high latitudes, migrate, most do not. Some, especially large eagles, are of economic importance, preying upon animals maintained by man (sheep, game birds, etc.), eating otherwise noxious carrion, or feeding upon pests of human crops (hares, rodents, game birds, locusts). The low population density of most species indicates that their economic importance is usually exaggerated. They do not feed upon man himself (except when he is dead, in the case of vultures), do not eat babies (as eagles have been accused of doing), and are not important disease vectors.

Falconiforms vary from tiny falconets (*Microhierax*), weighing 35 grams (1.2 ounces) or less, to huge vultures and eagles, with weights up to 14 kilograms (31 pounds), certainly the most formidable avian predators. Structure varies according to hunting methods, type of prey, and habitat. Carrion-eating vultures have broad, soaring wings and weak feet; speedy falcons have bullet-shaped bodies and long, pointed wings; manoeuvrable forest-living species, such as accipiters, some eagles, and forest falcons, have short, rounded wings and long, rounded or graduated tails. The beak is always hooked and is used to tear flesh from prey held in the feet; it may be further modified for special types of food, such as snails or bone marrow. The feet, which are the primary means of killing prey, are exceptionally strong, equipped with long, curved talons, except in the vultures, which feed on carrion. Sight and hearing are highly developed, but the sense of smell is usually poor or absent.

Distribution

Falconiforms occur worldwide, on all continents except Antarctica, and on many oceanic islands. They are found from Arctic to equatorial latitudes, in habitats ranging from almost stark desert to tropical forest, but are commonest and most varied in warm, relatively open country; many, especially in South America, have adapted to forest life. They are never abundant, except locally, where gregarious species may roost or feed together, and are usually dispersed at the rate of a pair to 15 square kilometres (six square miles) or more. Their numbers are not always density dependent on those of food animals, but, where food is abundant, raptor numbers may be limited by territorial behaviour.

NATURAL HISTORY

Life history. Large falconiforms are among the longest lived birds, as shown both by records in captivity and by the slow breeding rates observed in the wild. Where not persecuted, populations of large falconiforms are remarkably stable; breeding rates are presumably adjusted to the mean life span of individuals, taken from hatching. A species that averages two young annually per pair has a mean life span of one year; and one that rears one young every two years, a life span of four years.

Four stages of life cycle

The falconiform life cycle includes four stages, each of which varies mainly according to size—the larger the bird, the longer the stage. Latitude also affects these periods, with longer fledging periods in equatorial than in temperate latitudes for species of similar size.

The fledging period. The fledging period, from hatching to first flight, varies from 23 days in small accipiters to 120–130 days in large eagles and vultures and up to 150 in condors. In this stage the downy nestling develops to adult size (weight increases up to 40 times), grows feathers, and finally flies. It must be fed at first, but before its first flight it learns to tear up prey brought by the parents. Fledging periods are longer, in birds of a given size, in the tropics, probably associated with shorter summer day length.

The post-fledging period. The post-fledging period, in which the young remain near the nest and food is brought by the parents, ends when the young become independent, apparently often of their own accord, not, as in some other birds, in response to increasing parental aggression. They grow no larger, but their flight quills harden. These immature birds are not strong on the wing for three to eight weeks after the first flight. This phase varies from one to 11 months or even more, again mainly according to size, but also showing specific variation. In the crowned eagle, the post-fledging period is nine to 11 months, but in the related martial eagle it is much shorter.

The period as a juvenile. The period as a juvenile, an independent, sexually immature form, is distinguished by various immature and subadult plumages, often very different from the adult. The young raptor flies about, migrates if necessary, kills its own prey, and behaves like an adult except that it cannot breed. This period varies from about nine months in small species, sexually mature in their first year (sparrow hawks, kestrels, and some kites), to four or more years in large eagles and up to seven years in condors. It is usually a period of heavy mortality, particularly in migrant species of temperate climates.

The breeding adult life. The breeding adult life is characterized by mating, nesting, and the rearing of young. Most adults pair. The duration of this phase has been rarely recorded for wild individuals, but a peregrine falcon (*Falco peregrinus*) bred on the Sun-Life Building in Montreal, Quebec, for 16 years. Two female crowned eagles bred for nine and eight years and one male for 13 years.

Life spans of wild raptors may be calculated from banding (ringing) records, from observed changes of mate, and from the percentage of immatures in the population, especially if different ages are recognizable.

Life span

Recoveries of banded birds indicate that 50–75 percent of young raptors die before sexual maturity. The smaller falcons and accipiters apparently average less than three years' total life span, but some individuals may attain 10 to 12 years. Larger species live longer. Ospreys and buteos may average eight to 10 years but have higher mortality rates (75–80 percent in the first year). Life spans of more than 20 years have been recorded from both groups, but the longest known life span among wild falconiforms is that of a red kite of 26 years.

In certain large falconiforms, differences in plumage due to age and individual variation allow an observer to recognize individual birds. Since large eagles mate for life, it may be assumed that a change of mates indicates that one of the pair is dead. L.H. Brown has found that in crowned eagles (*Stephanoaetus coronatus*) such changes occur every six to seven years, giving a mean adult life of 12–14 years or 16–18 years when the four years prior to maturity are included.

In many species immature birds are easily distinguished from adults, and the percentage of immatures in the population can be related to assumed mortality before sexual maturity and to the number of years before adulthood, allowing an estimate of the average age of adults. Immatures comprise about 30 percent of the population in the bateleur (*Terathopius ecaudatus*), 18 percent in the lammergeier (*Gypaetus barbatus*), and 16–25 percent in the African fish eagle (*Haliaetus vocifer*). With about 75 percent loss during the period of immaturity and the acquisition of adult plumage at seven, five, and four years, respectively, the mean adult life span in these species may be estimated at 23, 27, and 18–24 years.

Such methods provide, at best, an informed estimate and the potential life span is rarely, if ever, attained in the wild state. In captivity many large raptors live for over 30 years, eagles and vultures for 40–50 years. A bateleur, at 55, is the oldest acceptable record; records of 60 or more years are poorly documented.

Behaviour. Falconiforms hunt by sight in daylight and normally roost on a perch, a tree branch, a rock ledge, or occasionally (harriers) on the ground in long grass. A few, including some falcons and sparrow hawks, are active at twilight; in fact, the bat hawk (*Machaerhamphus alcinus*) catches all its food in about half an hour at dusk. The only species actually recorded feeding at night are certain Indian vultures (in moonlight on tiger kills). Most raptors sleep all night, with head buried in the feathers of the back or hunched upon the shoulders.

General activity

Activity may begin soon after dawn with short "warming-up" flights, serious hunting beginning later. Some accipiters commence hunting at daybreak, when their avian prey is most active. Large, heavy species, however, do not usually start hunting or foraging until some time after dawn. Large vultures are unable to fly until updrafts are generated by thermal activity, but smaller species are able to fly soon after dawn. Certain specialized types with flapping flight, such as harriers, can fly equally easily at any time of the day. Bateleurs get on the wing early and fly most of the day, travelling perhaps 200–300 air miles on most days; however, very large species such as eagles and vultures probably fly for no more than four to six hours a day.

Before the day's flight, a raptor usually preens, casts, and defecates. Castings are regurgitated indigestible balls of fur, feathers, insect hard parts, etc. Preening is performed mainly with the bill, but falconiforms also scratch with their formidable talons. They frequently "rouse," fluffing out and shaking all their feathers.

Hunting and feeding

Many falconiforms hunt daily, others every few days, feeding in the interim on the remains of their kills. Killing frequency depends on appetite and the size of the animal killed; a kestrel consumes many grasshoppers per day, a crowned eagle one suni antelope (about four kilograms; eight pounds) every three to four days. Appetite and crop capacity permit ingestion of several days' ration at a meal, and larger species can go without food for long periods without ill effect. With rare recorded exceptions, falconiforms will not hunt unless hungry and do not kill wantonly. Many species hunt from perches, flying from one to another. Others hunt on the wing, soaring or hovering at up to 100 metres above the ground or water. One species, the secretary bird (*Sagittarius serpentarius*), is mainly terrestrial, and several others (chanting goshawks, honey buzzards, and the spotted eagle) walk about on the ground. The honey buzzard digs out wasps' nests. One species, the Egyptian vulture (*Neophron percnopterus*), hurls stones with its beak to break ostrich eggs.

In warm climates, hunting time may be short, one to two hours in 24; in temperate winters, when prey is scarce, it occupies up to five to seven hours. When not hunting, falconiforms loaf on perches or soar, sometimes at great height. Vultures or eagles encountered soaring very high up are not hunting, especially when their crops are full.

Roosting

Toward evening a falconiform may return to a regular roosting place or may settle for the night where it is.

Drawing by R. Keane

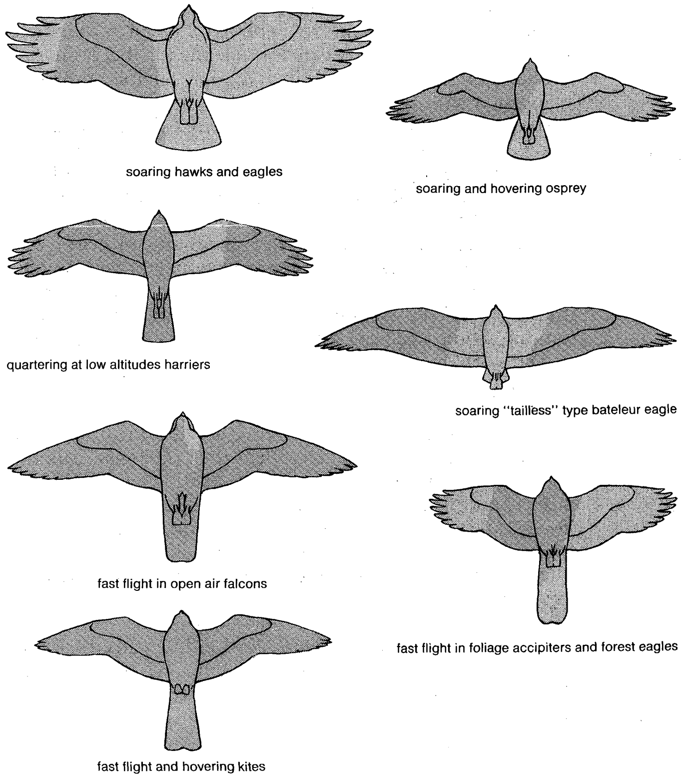


Figure 22: Modifications for specialized types of flight among falconiforms.

Vultures often return nightly up to 100 miles to regular roosting cliffs or trees. In many less active species, the roost is in the same general area as the nest. Members of a pair separated all day may rejoin at roosting time, and gregarious species (vultures, kites, and some others, such as the red-footed falcon) are most gregarious at roosts. Prior to settling to sleep, the bird usually preens again, but activity is reduced once the bird has settled on the perch for the night. Finally, when light fails, the raptor falls asleep on its perch and remains thus without moving until the following morning. If disturbed at night, however, it can fly well enough to save itself but will not travel any distance.

Locomotion. All falconiforms fly well, some excellently. Few hunt on the ground, but all can walk or hop. With one exception—the secretary bird—flight is the most important method of locomotion. Modes of flight include flapping, gliding, soaring, or diving (stooping) and are used to locate and kill prey, move from perch to perch, or migrate. Adaptations for swift or slow flight, mainly associated with habitat or type of prey, are varied.

All species can use flapping flight, but it is laboured in heavy species (vultures, condors). Most use it only to get under way or to move from perch to perch, thereafter gliding, hovering, or soaring. Harriers sustain flapping flight, interspersed with short glides, for long periods as they hunt small animals in open, grassy country. This also assists them to migrate across open water, usually avoided by soaring species because of the lack of thermal updrafts. Accipiters accelerate in pursuit by rapid flapping flight, and many other species travel short distances by flapping alternated with gliding, especially when air currents are lacking.

All of the large species are adapted for soaring flight and even accipiters, swift falcons, and the secretary bird can soar well. A typical soaring wing is rather long and moderately broad; *i.e.,* of low aspect ratio. High-aspect-ratio wings (long and extremely narrow), found in seabirds, would presumably be inconvenient to falconiforms, which often require manoeuvrability in tight spaces. The wing usually has the outer four to six primaries more or less emarginated on one or both webs, forming at the spread wing tip slots that reduce turbulence, and permitting each feather to bend as an individual airfoil under a differential load, thereby increasing lift. Such a wing is typical of eagles, buzzards, vultures, and kites. In a few species (Indian black eagle, African harrier hawk), the primaries are long, soft, slightly emarginated, and flexible, probably associated with slow flight.

Soaring on updrafts

The ability to soar and circle in thermals is controlled by wing loading (the ratio of weight to wing area). The higher the wing loading, the larger the turning circle and the larger the thermal "bubble" required for soaring to gain height. Smaller species (*e.g.,* the black kite), with low wing loadings, can utilize smaller thermals than can heavy vultures. Wing loading increases proportionately with weight (as the cube of body size) while wing area increases as the square. The general rule is varied by specific adaptations. Some species (lammergeier, harriers) have much lower wing loadings than others of similar weight and have more buoyant, easy flight. In soaring flight, forward velocity is attained by gliding, with the aid of gravity; the bird mounts on rising air. Increasing the angle of glide increases speed until, in a vertical dive (falcons), it is believed to attain 160–240 kilometres (100–150 miles) per hour. Some species (bateleur, lammergeier) are specialized for gliding, almost without wing flapping. Bateleurs glide at an airspeed of 55–85 kilometres per hour for much of the day; they have long wings and very short tails, steering by sideways tilting, or canting. Lammergeiers glide rather slowly, assisted by long, diamond-shaped tails.

Falcons, especially the large bird-killing species (peregrine, gyrfalcon, lanner, etc.), have heavy, bullet-shaped bodies and long, pointed wings; their high wing loading provides high diving speed. They often attack in the air and in a dive can overtake species that would easily evade them in straight flapping flight. Few eagles are as swift, but many combine distinguished soaring and gliding ability with swiftness and agility in attack upon prey.

Walking and hopping

On the ground falconiforms progress by walking or hopping; in especially large vultures hopping is elaborated into bounding threat displays. On a branch they move sideways by sidling or by walking "hand over hand" (vulturine fish eagle, harrier hawk). On the ground eagles walk slowly and deliberately. African harrier hawks and South American crane hawks have long, slim legs that can bend somewhat backward at the tarsal joint, permitting the attainment of peculiar situations and enabling the birds to probe cavities.

The secretary bird is the only primarily terrestrial species. It flies only to move from place to place or in display but nevertheless has been seen soaring at 4,000 metres. It walks with a rapid, steady pace, head jerking to and fro, varying this regular progress with bouts of quick stamping steps, to flush possible prey. It has short toes (like the gruiform bustards) and long, cranelike legs, suited for rapid walking over long distances. It roosts in trees but descends to the ground to hunt soon after daybreak.

Reproduction. Many falconiforms pair for life; others, notably migrant species, may pair anew each year. Occasionally, immatures pair with adults, but usually only sexually mature individuals attempt to breed. Immatures, however, often perform typical display movements and a pair bond may be formed between an adult and an immature, but the latter may be replaced by an adult if one should appear.

The breeding pattern follows the usual sequence of nuptial display, nest building, incubation, fledging period, and post-fledging period. The length of the breeding period varies from about three months in the smallest species to over 15 months in condors, crowned eagles, and probably harpy eagles. No species regularly produces two broods in a season, but second broods have been recorded, even in some large species. Some nomadic species, such as the elanine kites, may breed more often in response to, for instance, high local rodent populations.

Nuptial displays are often spectacular, sometimes inconspicuous. The common forms are: calling from a perch; calling from soaring flight; undulating flight, diving, and swooping up, often with loud calling; mutual displays, in which the male dives with lowered feet at the female, who turns over and raises her claws to his; rarely, cartwheeling, in which the pair lock feet and descend in whirling flight (*e.g.,* some eagles and kites). Most species perform several types of display, but species inhabiting woodland or forest display more often and more vociferously than related species of open country.

Nesting

Most species build nests on trees, ledges of cliffs, or, rarely, on the ground. All Accipitridae, Pandionidae, Sagittariidae, and the caracaras construct nests, usually of sticks. The Cathartidae and the remaining Falconidae (forest falcons, falconets, and true falcons) do not make nests but use a hollow tree, another bird's nest, or a scrape on a ledge. Constructed nests vary from small structures of twigs and vines (*Aviceda*) to enormous, annually augmented piles of sticks (many eagles). Small species usually build a new nest annually; large species tend to return to, and repair, an old nest. Imminent occupation is usually indicated by fresh green material, which continues to be added until late in the fledging period but does not necessarily mean that eggs will be laid. The same site, but not always the same nest or ledge, may be occupied for several years, generations, or centuries. Some large eagles use the same nest for 50 years or more, others haphazardly use several alternates.

Many species lay only one egg; few lay more than four. The largest clutches are laid by small falcons and accipiters of temperate latitudes and by ground-nesting harriers. Clutches of related species are smaller in the tropics than in high latitudes. The eggs are laid at two- to four-day intervals, incubation beginning with the first egg. Falconiform eggs are rather large for the size of the birds and are usually rounded ovals, white or greenish, more or less heavily spotted, occasionally completely covered with brown or red-brown pigment. In Carthartae the colour of the inside of the shell is yellowish white; in Accipitridae, Sagittariidae, and Pandionidae it is green or blue green; and in Falconidae, buff or reddish. Incubation, usually by the female alone (the male sometimes takes a share), occupies from 28 days in small falcons to 50 days or more in large vultures. Buzzards (*Buteo*) incubate for 30–35 days, though some are larger than small eagles with incubation periods of 45 days.

The eggs hatch in the order in which they were laid. By the time the last egg has hatched, the first-hatched chick may have tripled its weight and may kill its sibling, either directly or through competition for food. Newly hatched chicks are covered in natal down, later replaced by a thicker, woollier coat; at this stage parental brooding is reduced by day. Feathers erupt during the fledging period and almost cover the nestling at midpoint. At this stage most young learn to tear up their own prey.

Fledging

Parental attention is steadily reduced, from continuous brooding day and night to leaving the feathered young entirely alone except for brief feeding visits. The young make their first flights unaided by parental coaxing. The adult roles are usually clear-cut. In the early fledging period, the male kills for himself, his mate, and the brood doubling or trebling his previous killing rate. When partly feathered young can tear up their prey, the female no longer stays near the nest but hunts and thereafter brings more prey than the male.

After their first flight, the young remain near the nest for several weeks or months. They are provided with food by the parents but may also learn to kill. The longest known post-fledging periods are in crowned eagles and condors; in the crowned eagle the young remains dependent for nine to 11 months, and this species normally breeds only every second year.

Especially in large species, breeding does not necessarily occur annually. The number of young reared per pair per year varies from 2.5 to three in small falcons and sparrow hawks of temperate climates to 0.4 or even less in some tropical eagles and vultures. Breeding success, like clutch size, tends to be lower in related species in the tropics than in temperate zones. Low breeding success may fluctuate with adverse winters, poor food supplies, or even territorial behaviour in crowded habitats.

Ecology. Falconiforms prey on small animals or eat carrion; the vulturine fish eagle, however, feeds mainly on oil palm fruits.

Food supplies exert a general influence over populations, but it also seems that when food is superabundant the population does not rise proportionately. In the Scottish golden eagle, the home range of a pair is about 4,400 hectares (11,000 acres), irrespective of food abundance. A breeding pair of raptors apparently maintains a territory containing much more food than it can possibly utilize. In some cases such home ranges are maintained by aggressive behaviour; in others the precise method is obscure. In colonial vultures no home range is maintained, and the only defended territory is the few yards round the nest.

Food requirements

From a combination of factors a reasonably accurate estimate of the minimum daily food need for a raptor can be made. The amount of kills wasted (varying from less than 5 to 20 percent by weight, according to type of prey taken) is added to the minimum daily food demand, to determine the amount that must be killed to maintain a pair and their young for one year. For example, if a falconiform weighs 1,200 grams, has a mean food requirement of 8 percent of bodyweight, eats 96 grams a day and 36 kilograms each year, and feeds on birds, involving waste of about 15 percent of all kills, it must kill 41 kilograms of prey each year to fulfill its needs.

Individuals of one species do not usually tolerate others of the same species in their home breeding range. The same area, however, may support several other species, preferring different types of prey and so not competing with one another. In an area of varied habitat, up to 20 different species may occur (even more with migrants), exerting a much greater and more complicated, but still broadly quantifiable, pressure on populations of available prey. Closely related groups may show ecological separation. For example, some African vultures feed in groups upon the soft flesh and intestines of preys, while others tear at sinews and hide or snatch meat from other species. Smaller vultures pick up scraps or use their slender beaks to reach cavities inaccessible to birds with heavier beaks. All these birds may feed together at the same carcass, but without competing for food, though individuals of each species maintain a dominance hierarchy by aggressive movements to each other.

In the Arctic only the gyrfalcon does not regularly migrate but subsists on wintering ptarmigan. In summer several other species arrive, but so do many species of prey, and lemmings multiply, greatly increasing food supplies. In African savannas and southern Indian rice fields, resident falconiform populations are greatly augmented in winter by migrant populations of buzzards, kites, harriers, and small falcons. This sharply increases the total predator pressure on prey populations, but it does not appear that this influx of migrants has any adverse effect on the resident populations, which may be breeding while the migrants are present.

In most cases it appears that falconiforms have a slight or negligible effect on prey populations. The annual needs of resident species may vary from $^1/_{20}$ to $^1/_{1,000}$ or even less of available prey. In migrant winter populations, the effect, locally, may be much greater than in resident breeding populations. The availability or visibility of the prey also plays a part. Among the food of the European sparrow hawk (*Accipiter nisus*), conspicuous or gregarious birds of open ground are more often killed than inconspicuous skulking species. Prey risk (the likelihood of an individual's being killed) is increased by low prey numbers, reduced cover, and behaviour that exposes the animal. The total effect of falconiforms on their prey is compounded of their own behaviour and that of the animals they feed upon.

IMPORTANCE TO MAN

Beneficial effects

Most falconiforms are directly beneficial to man—feeding on pests or carrion—or neutral—feeding on animals and birds that do not directly affect man. A few species, usually large, spectacular, and uncommon (larger eagles, falcons, and accipiters), may prey on man's domestic stock or poultry or on game birds or fish. In East African grasslands about 86 percent of falconiforms are beneficial, 11 percent neutral, and only 3 percent potentially harmful. Such figures vary from one region to another, but potentially harmful species are always in the minority. It is the more regrettable that, because of ignorance and unreasoning prejudice (especially in the more advanced areas such as North America, Europe, and Australia), beneficial, neutral, and potentially harmless species are all indiscriminately destroyed, even rendered extinct locally.

Detailed quantitative studies invariably have shown that the harm done to man's interests by any falconiform has been grossly exaggerated; even the supposedly harmful species often kill other potentially harmful species (*e.g.*, crows, rats, mustelids). The predation effect of any species is relatively small in relation to the total number of potential prey in any home range. Moreover, even in potentially harmful species, few animals of potential or actual interest to man are taken. In the African fish eagle and Scottish golden eagle, the population may take a fraction of 1 percent of the human fish catch or of the lambs available in an average home range. In North America and Australia, where large eagles have been destroyed on a greater scale than elsewhere, persecution is not based on even rudimentary assessment of the eagles' diet and appetite. Damage to human interests is worst under poor management conditions; for instance, when unfed poultry are on free range or weak lambs live on overgrazed pasture. In rare cases, such as where hatchery trout or reared pheasants are unnaturally concentrated, individual hawks or eagles may have to be controlled when damage is proved; but this is no excuse for wholesale persecution of a species.

Entirely beneficial species include vultures, elanine kites, kestrels and other small insectivorous falcons, most buzzards (*Buteo* and related species), and many small eagles. Neutral species, such as snake eagles and chanting goshawks, can safely be ignored. Objective analysis of the predation situation indicates that diurnal raptors deserve protection except where extensive damage can be proved. Extensive damage is frequently claimed without proof; for instance, in the southwestern United States, where the golden eagle has been persecuted for alleged damage to sheep, improbable in any event and impossible on the scale claimed.

A few falconiform birds have been trained to secure game animals for man, either for sport or to provide food.

Need for protection

Falconiforms need no protection except from man. Conservation problems, therefore, mean conserving raptors living in or passing through inhabited areas. Undoubtedly the main difficulty is ignorant prejudice against falconiforms by educated, civilized man, which is seldom found, much less effectively exercised, in more primitive human communities. A recent threat, since 1950, is the persistent effect of highly toxic agricultural chemicals (DDT, dieldrin, etc.)—again, the product of civilized man. Positive conservation of falconiforms must include the control of habitat destruction; of active persecution by shooting, trapping, poisoning, and collecting of eggs of threatened species; and of the widespread toxic effects of persistent agricultural chemicals.

Agricultural communities destroy some types of habitat, especially forests, but create others in the process. In Africa, for instance, some forest falconiforms may be eliminated, but much larger populations of adaptable species such as black kites, hooded vultures, and long-crested eagles may move in. The effect of agriculture is not always detrimental, especially when a mixed landscape of forest and open country results. Before the advent of chlorinated hydrocarbons, the peregrine was most numerous in Britain in habitats close to agricultural areas with high populations of avian prey. A sparse population of rural cultivators may have no appreciable effect on a mixed population of falconiforms over long periods of time.

Overcollecting for zoos and museums, and for falconry, has severely affected some species. The behaviour of the few individuals who collect the specimens or eggs of threatened species for commercial purposes must be controlled. In a few instances, chiefly of island species such as the Philippine monkey-eating eagle, this is urgent.

Effect of pesticides

Toxic agricultural chemicals are the most difficult of conservation problems because of their subtle and widespread effects (affecting even migrant populations of some hawks that breed in uninhabited areas) and the benefits they confer on human food supplies. The best hope, other than new developments of less toxic substances, is realization of the probably serious overall effect of these long-lasting poisons in contamination of the whole environment. If, as appears likely, the chemicals eventually will adversely affect man himself, control of the use of these chemicals will follow; and the falconiforms will benefit indirectly.

FORM AND FUNCTION

The main distinguishing characters of falconiforms are: the hooked beak, used for tearing flesh; taloned feet, used for piercing, grasping, and killing prey; and large eyes, with very acute vision. The hooked beak and taloned feet are shared with the unrelated owls, which show similar or parallel evolution. In the falconiforms themselves, parallel evolution for scavenging is shown by the New World vultures (Cathartidae) and the Old World vultures (Accipitridae).

Parallel evolution

Size and weight. Most falconiforms are large. Extremes are the Andean condor (10 to 14 kilograms, wingspan three to 3.35 metres) and the tiny eastern falconets (*Microhierax* species; 35 grams or less). Many species range from 400–2,000 grams in weight. Females are normally 20–100 percent larger than males, especially in bird-killing falcons and accipiters. In vultures, however, the male is normally 10–15 percent larger.

Plumage and molt. Falconiforms are bulky, heavily feathered birds, lightweight in relation to their apparent size. The first plumage is usually white or gray down, the prepennae. The preplumulae, or second down, grows through the first after days or weeks and is itself superseded by feathers erupting from the prepennae follicles. The latest feathers to develop fully are wing and tail quills, which are large, strong, and often especially adapted.

The immature plumage, presumably representing a primitive type, usually differs markedly from the adult. Adult dress is acquired by a series of molts with, in large species, several intermediate or subadult stages. Immatures usually are brown and streaked or spotted; adults may be more brightly coloured. The sexes usually are alike in plumage. In some island species—*e.g.*, the Madagascar cuckoo falcon (*Aviceda*)—the plumage type found in the immature persists in adult life.

Plumage is replaced by a molt lasting four months to several years, slower and more irregular in larger species. Wing quills are molted in definite sequence. In many Accipitridae molt begins with primary feather number 1 (innermost) and proceeds in regular sequence to number 10 (outermost). In falcons it begins at primary feather 4 and proceeds both ways, outward and inward. Large species tend to replace quills irregularly, perhaps because their greater wing loading would make the lack of several adjacent quills disadvantageous.

Vision, hearing, and smell. The eyes are globose and move little in their sockets. To see behind it, a falconiform must rotate its head. The focal length is adjustable by muscles controlling the curvature of the lens. Forward binocular vision is through 35–50° of arc. The proverbially high resolving power of hawks' eyes depends partly on a large image on the retina, partly on the concentration of rods and cones. There are two foveae (areas of high visual acuity), one laterally directed for monocular vision, the other forward, for binocular vision; in each fovea the visual cells are still more concentrated, providing resolving power up to eight times that of the human eye. There is a screenlike pecten, which may cast a lattice-like shadow on the retina, permitting good perception of shape of moving objects. The iris in many adults is yellow, red, or orange.

Hearing is good, but not especially acute. The ear apertures are large and in harriers and forest falcons are above average size and surrounded by especially modified feathers, forming partial facial ruffs. The forest falcons live in dense woodlands and are seminocturnal, and the harriers hunt small mammals in long grass, both situations where unusually acute hearing would be advantageous.

Except in the New World vultures, which occasionally use their sense of smell to locate food, the olfactory chambers are poorly developed. The sense of smell is normally rudimentary and less important than sight.

Feet and bill. These appendages comprise the main killing and feeding adaptations distinguishing birds of prey. The exact structure of the bill varies according to prey eaten (see Figure 23). Falconidae and some insectivorous kites have notches or teeth on the cutting edge of the beak. In falcons these assist in breaking the necks of prey, but their purpose in kites is obscure. In Old World vultures the bills vary, permitting ecological separation while at food.

Killing of prey

Prey is normally killed with the feet. Three toes are directed forward, one behind. The hind toe is usually heavier and longer taloned than the others. In the osprey the outer toe is reversible, for more effective handling of fish. Fish-eating falconiforms have sharp spicules on the soles of the toes to grasp slippery prey. The feet may vary from long, slim toed, needle taloned (bird killing, as in accipiters) to short, thick toed (snake grasping, *Circaetus*) or heavy toed, with thick, strong talons, capable of paralyzing medium-sized mammals. In Old and New World vultures, which seldom or never kill, the feet lack long sharp talons.

EVOLUTION AND PALEONTOLOGY

Falconiforms have no obvious evolutionary links with other birds. Currently they are placed between ducks, on the one hand, and game birds on the other; but they bear no clear resemblance to either, while fossil evidence does not indicate intermediate links. The most obvious physical specializations of falconiforms—the cutting, tearing bill and taloned feet—do not indicate close relationships with owls but are due to similar trends in evolution, though some recent anatomical work indicates that owls and the family Falconidae may be more closely related than was at one time thought.

Fossil falconiforms

Rather few fossils of falconiforms have been found, and those that have may repay re-investigation. A generalized type is known from the Eocene (50,000,000 years ago). The oldest known raptorial bird, *Lithornis,* of the upper

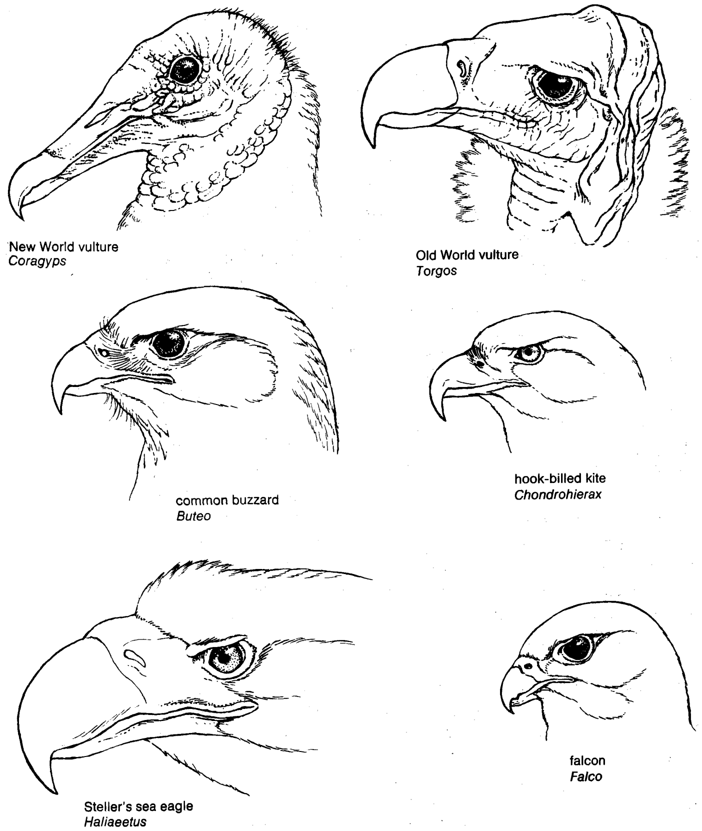


Figure 23: *Heads of some falconiforms.*
A New World vulture, *Coragyps,* with a weak beak for carrion eating; an Old World vulture, *Torgos,* with a stronger beak for tearing at larger animals; a buzzard, *Buteo,* with a simple raptorial beak for killing and eating small mammals; a hook-billed kite, *Chondrohierax,* with a strongly hooked beak for eating snails; a sea eagle, *Haliaeetus,* with a deep, narrow beak of uncertain adaptive value; a falcon, *Falco,* with a toothed beak for shearing and plucking feathers.
Drawing by R. Keane

Paleocene, may have been a cathartid vulture. The Cathartae may have evolved in the Old World, dying out there and surviving only in the New World. Fossil New World vultures include a large terrestrial species, *Neocathartes grallator,* and a huge vulture, *Teratornis merriami,* from the Rancho La Brea tar pits in California. Because of their apparently ancient origin, cathartids may be regarded as primitive survivors.

The secretary bird is thought by some to be a descendant of formidable terrestrial predators, *Brontornis, Mesembryornis,* and *Diatryma.* These also may have evolved into the South American Cariamidae, which accordingly may be more closely related to the secretary bird than are the latter to other falconiforms.

Such evidence as there is indicates that the falconiforms diverged from their ancestral types 20,000,000 or more years ago, and that relationships with other groups have since become obscured. Certain evidence suggests that the order is polyphyletic in origin; however, their Mallophaga (feather lice) are similar throughout, and this question must remain unanswered until extensive additional information is available.

Functional specialization

Among present-day falconiforms adaptive radiation has led to the development of certain climax types in apparently not very closely related birds. For instance, in Australia, buzzards (*Buteo*) are absent, but certain large kites have evolved to fit this ecological niche. In South America the buzzard-like harpy eagle represents a climax of one line of evolution, filled in the Old World by large booted eagles such as *Stephanoaetus.* There has been much adaptive specialization for particular function in unrelated species; for instance, in the bullet-like shape of falcons, in the spiculed feet of ospreys and certain eagles, and, perhaps most striking, in the auditory specialization of the harriers and the forest falcons.

CLASSIFICATION

No entirely satisfactory system of classification of the order Falconiformes has been agreed upon. Two rather different systems have been proposed, and much work is needed on several outstanding problems.

Distinguishing taxonomic features. The main features of falconiforms include the hooked "raptorial" bill; the basal cere, or bare skin, covering the nostrils; the powerful feet with hooked claws; sustained powers of flight and carnivorous habits; and difference in size between the sexes. Some suborders or families have additional characters: the Cathartae have pervious nostrils (*i.e.,* incomplete partition between nostrils) and a rudimentary hind toe; Pandionidae, a reversible outer toe; Sagittariidae, long legs with short, blunt toes; and Falconidae, toothed beak, sharply pointed wings, and a noncentrifugal wing molt.

On the basis of inside eggshell colour, the order may be divided into three rather distinct groupings or suborders: the Cathartae, the Accipitres (with the secretary bird and osprey), and the Falcones. Biochemical analysis of egg-white protein tends to support these groupings and suggest that all are rather more closely allied to one another than to other groups of birds, even though there are some similarities with apparently unrelated orders, such as Procellariiformes. Egg-white proteins also indicate that cathartids are all closely interrelated and that they are related to Accipitridae. The Accipitridae are all rather closely related, though the egg-white patterns of Old World vultures and some kites indicate divergence. The egg-white patterns of Falconidae indicate close relationship within this family, but less close affinities with other falconiforms.

Annotated classification. The classification of L.H.

Drawing by R. Keane

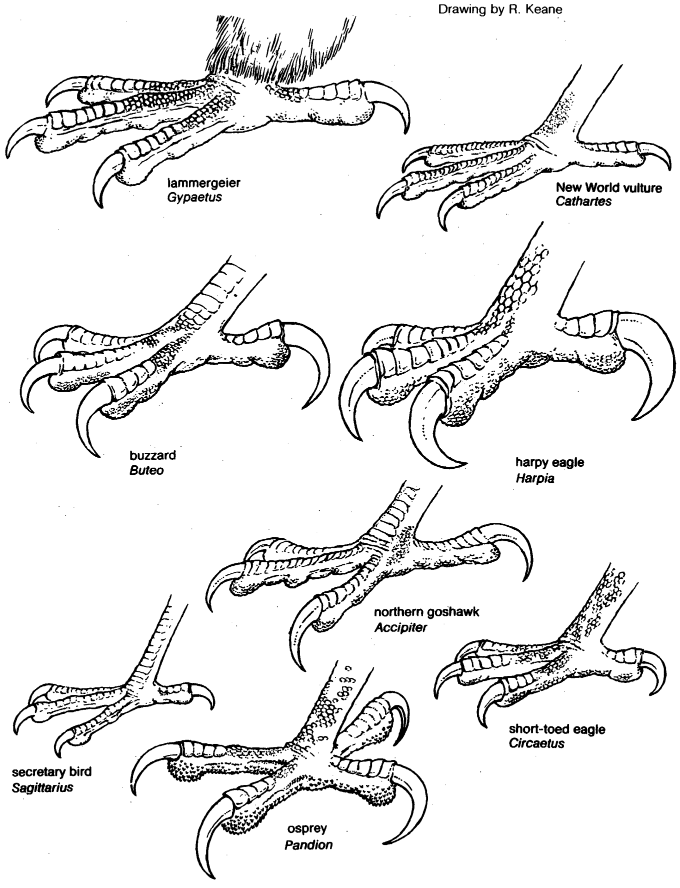


Figure 24: *Specialization of the feet of falconiforms.*
The moderately powerful foot of the lammergeier (*Gypaetus*); the weak foot of a New World vulture (*Cathartes*); the generalized raptorial foot of a buzzard (*Buteo*); the extremely powerful foot of the harpy (*Harpia*); the foot of a bird-catching accipiter (*Accipiter*), with long toes and talons; the fish-gripping foot of the osprey (*Pandion*), with reversible outer toe and rough spicules on the soles; the foot of the short-toed eagle (*Circaetus*) for gripping snake; the foot of the secretary bird (*Sagitarius*) adapted for walking.

Brown and D. Amadon, 1968, which is given below, was based upon that in J.L. Peters' *Check-List of Birds of the World,* modified as necessary in the light of modern knowledge. The main divergent view has been that of E. Stresemann, which was based largely on the sequence of wing molt and departed strongly from the order of Peters and from Brown and Amadon. Those groups indicated below by a dagger (†) are known only from fossils.

ORDER: FALCONIFORMES

Suborder Cathartae (New World vultures)

Nearctic to Neotropical. Seven existing species; fossils dating back perhaps to Eocene.

†*Superfamily Neocathartoidea*

One family, Neocathartidae. Fossil Nearctic, including *Neocathartes grallator,* differed from other Cathartae in having terrestrial habits.

Superfamily Cathartoidea

Pleistocene fossil and present-day species, Nearctic and Neotropical, perhaps formerly Palearctic.

Family Cathartidae (New World vultures). Seven present-day Nearctic and Neotropical species and Pleistocene, Nearctic, and neotropical fossils. Pervious nostrils, rudimentary hind toe, large olfactory chambers. Large or medium sized (1.5–14 kg; 1.2–3 m wingspan); do not build nests; inside egg colour yellowish.

†*Family Teratornithidae* (teratorns). Very large-sized fossil Pleistocene species, including *Teratornis merriami* (estimated about 22.5 kg; 4.3 m wingspan) and still larger species.

Suborder Accipitres

Superfamily Accipitroidea

There are 207 worldwide existing species, small to large, and many fossils.

Family Pandionidae (osprey). One species, *Pandion haliaetus,* nearly worldwide, present-day; large, weight 900–1,500 g; wingspan 135–175 cm; resembles some kites in similar sternum and absence of bony eye shield; reversible outer toe, with talons rounded, not grooved; green inside eggshell; squirt droppings; littoral; builds own nest.

Family Accipitridae (kites, hawks, buzzards, harriers, eagles, Old World vultures). There are 206 species, worldwide, but some subgroups—*e.g.,* Old World vultures—confined to Europe, Asia, and Africa. Small to very large (100 g to about 10 kg); many adaptive radiations; forest or open country. In all, inside eggshell green; strong hooked talons in all but Old World vultures; squirt droppings; build own nests.

Superfamily Sagittaroidea

Family Sagittariidae (secretary bird). One species, *Sagittarius serpentarius,* central Africa; present-day, with some doubtful fossils. Large (weight 3.4–3.8 kg; wingspan 1.5–2 m); terrestrial, long legged, short toed; in nesting and display resembles Accipitridae. Inside eggshell green but certain other characters indicate affinity with South American Cariamidae (order Gruiformes). Usually accepted as falconiform but may not be.

Suborder Falcones

There are 61 worldwide species, present-day and fossils. No superfamily designations; the order comprises only 1 family.

Family Falconidae (caracaras, milvagos, forest falcons, falconets, and true falcons). Some genera (*Polyborus, Daptrius, Herpetotheres, Milvago, Micrastur*) confined to New World, and others (*Microhierax*) only Oriental; very small to medium sized (35–1,800 g). Resemble Accipitridae in hooked beaks, talons, sexual dimorphism, etc.; differ in molt sequence, some anatomical characters, behaviour (droppings fall below perch; head bobbing in some; in all but caracaras no nest made); inside eggshell reddish.

Critical appraisal. Stresemann's classification of present-day falconiforms differed from the one given above in being based almost exclusively on the order of wing molt, classified in three ways: (1) beginning with primary number 4 and proceeding inward and outward from there; (2) "descendant," beginning with primary number 1 and working outward to the wing tip; (3) "irregular," with no clear sequence. The third category can again be subdivided into types that molt "descendant" (number 2 above) in juvenile molt and irregularly in adult plumage, and others that molt irregularly in all stages of plumage.

On this basis the order Accipitriformes of Stresemann (*cf.* Falconiformes of most authors) was divided into three suborders, Cathartae, Accipitres, and Sagittarii. The suborder Cathartae contained the New World vultures (*e.g.,* Cathartae of Brown and Amadon); the Accipitres included all those species included by Brown and Amadon in their suborders Accipitres and Falcones, with the exception of the secretary bird, Sagittariidae. Stresemann places the secretary bird alone in a separate suborder, Sagittarii. He thus included all but eight species of present-day falconiforms in the suborder Accipitres. Stresemann divided his Accipitres into two families: Accipitridae, molting "descendant" (number 2 above) or irregularly; and Falconidae, molting from primary 4 (as in number 1 above). Each of these families was divided into a number of subfamilies, and it was in the arrangement and order of these subfamilies that Stresemann's order of classification differed most strongly from the other current systems.

The relative merits of these two classifications are still open to discussion. That of Stresemann has been criticized on the ground that virtually all of the birds that molt irregularly are either large or very large. This applies to all his Cathartae; to all or most of his subfamilies Gypaetinae, Aegypiinae, Polyboroidinae, Circaetinae, Haliaeetinae, Machaerhamphinae, and Pandioninae of his family Accipitridae; to the larger and heavier members of his subfamily Accipitrinae; and to the secretary bird. No member of his Falconidae molts irregularly, but then none, except some caracaras and perhaps the gyrfalcon, can be called large and heavy. It appears at least as probable that the order of wing molt is as related to weight, size, and wing loading as to phyletic relationships. No one, however, questions the obvious differences between falcons and hawks, which are more strongly stressed by Brown and Amadon in their suborder Falcones than by Stresemann, who gives the falcons only family rank.

Controversies of evolutionary ranking

Modern birds usually are classified by the principle of placing the most primitive first and leading successively to the most highly evolved. Brown and Amadon suggested that Stresemann's ordering of subfamilies in his Accipitridae reversed that order by beginning with the rather specialized Old World vultures and ending with the sea eagles, kites, and osprey, which they regard as relatively primitive falconiforms. The position of the osprey remains controversial. Comparison of egg-white proteins indicates that the Old World vultures are closer to the typical eagles and buzzards than to the snake eagles and harriers. Stresemann concluded his large subfamily Accipitrinae with the buzzard-like harpy and monkey-eating eagles (*Harpia* and *Pithecophaga*), which Brown and Amadon considered less highly evolved than the booted eagles with feathered toes, such as *Spizaetus* and *Stephanoaetus.*

The whole subject of the classification of the Falconiformes is, in several respects, in need of thorough critical reappraisal and review. The best of the existing classifications is a compromise that still leaves many questions of precise relationships unanswered. (L.H.B.)

Galliformes (grouse, pheasants, quail, turkeys)

The bird order Galliformes, or gallinaceous (*i.e.,* fowl-like or chicken-like) birds, includes about 240 species, of which the best known are the turkeys (family Meleagrididae); chickens, quail, partridge, pheasant and peacock (Phasianidae); guinea fowl (Numididae); and grouse (Tetraonidae). Lesser known members of the order are the megapodes and the chachalacas, guans, and curassows. Although the hoatzin is here treated with the Galliformes, evidence from studies of egg-white proteins indicate that it should be assigned to the subfamily Crotaphaginae of the family Cuculidae (order Cuculiformes).

GENERAL FEATURES

Size range and diversity of structure. Most galliforms are medium-sized birds, from the size of a pigeon to that of a domestic chicken, 40 to 60 centimetres (16 to 24 inches) long and 500 to 2,500 grams (one to five pounds) in weight. The smallest members of the order are the sparrow-sized painted quail (*Excalfactoria*), about 13 centimetres (five inches) long and about 45 grams ($1\frac{1}{2}$ ounces) in weight. The heaviest galliform is the common, or wild, turkey (*Meleagris gallopavo*), wild specimens of which may weigh up to 11 kilograms (about 24 pounds);

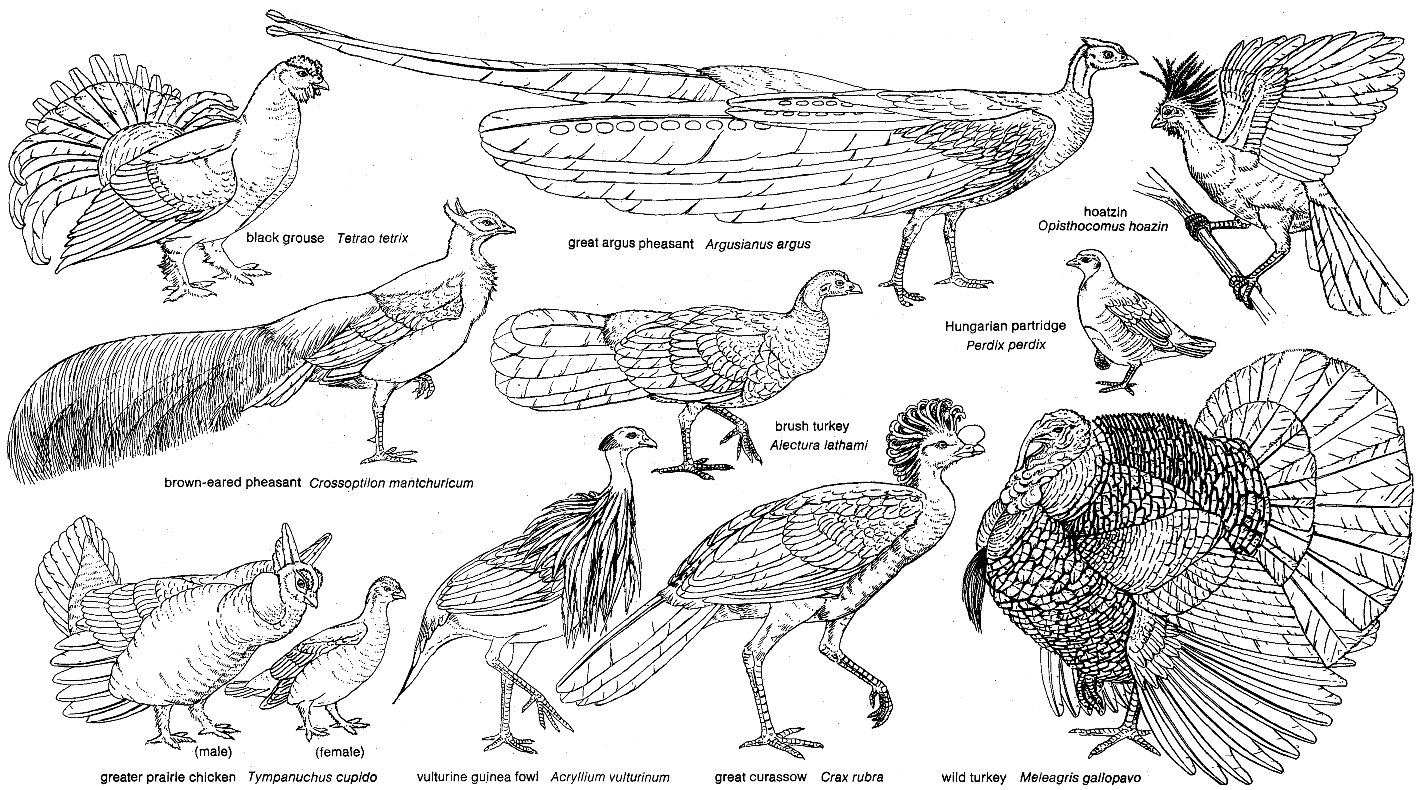


Figure 25: Body plans of galliform birds.

the longest is the argus pheasant (*Argusianus argus*), the male of which reaches two metres (79 inches) in breeding plumage, including wing and tail feathers, whose length exceeds one metre.

The majority of gallinaceous birds are heavy bodied, with short, rounded wings and strong, four-toed feet, adapted for life on the ground; a few, such as cracids and the hoatzin, live mainly in trees. The bill is short and slightly downcurved. The flight is fast but rarely sustained for long distances, most being sedentary, nonmigratory birds.

Distribution. Some members of the order are found in virtually every habitat in subarctic, temperate, and tropical regions of the larger landmasses, and a few species (ptarmigan, *Lagopus*) live within the Arctic Circle. The phasianids, the 178 species of which comprise by far the largest family, have nearly the distribution of the order. The cracids, with about 34 species, are restricted to tropical woodlands of Central and South America. The 18 species of grouse are found in northern temperate and Arctic regions of both hemispheres. The remaining groups are more restricted in distribution. The 10 species of megapodes inhabit forests from the East Indies east to the Fiji Islands and south to central Australia. Guinea fowl are restricted to Africa south of the Sahara but have been widely introduced on other continents. The two turkeys are native to North and Central America; the common, or wild, turkey in temperate woodlands of eastern United States and Mexico; and the ocellated turkey (*Agriocharis ocellata*) in Guatemala, British Honduras, and the Yucatán Peninsula of Mexico. The hoatzin (*Opisthocomus hoazin*) inhabits wooded river edges in northern and eastern South America.

IMPORTANCE TO MAN

Origin of the domestic chicken

Gallinaceous birds are unquestionably the most important avian group from the human standpoint. The chicken (*Gallus domesticus*) was domesticated in southern Asia at least 4,000 years ago from a parental stock of one or more species of jungle fowl (principally *G. gallus*). Selective breeding has produced well over 100 varieties, most of which are specialized for the production of either eggs or meat. In North America and to a lesser extent in northern Europe, the turkey is raised in numbers second only to those of the chicken. Domestic breeding of the turkey has been primarily for rapid growth and high market weight, the birds being raised solely for meat. Mature toms (males) may reach 23 kilograms (about 51 pounds). In many parts of the world, guinea fowl are an important barnyard species and are valued for the readiness with which they give alarm at the approach of a predator. (For a complete account of domestic gallinaceous birds, see FARMING: *Livestock and poultry farming.*)

Galliform birds constitute the large majority of land game birds, in number both of species and of individuals. Populations in North America and in western Europe are often carefully managed through habitat manipulation, supplemental feeding, and artificial rearing to ensure maximum hunting yields. The common pheasant (*Phasianus colchicus*, called the ring-necked pheasant in the United States) was introduced in North America late in the 19th century and has become widely established. Several species of quail (especially those of the genera *Coturnix* and *Colinus*) and partridge (*Perdix* and *Alectoris*) provide much sport hunting. A widespread tendency in game-bird management has been the introduction of species to areas where they have not been found previously.

The North American wild turkey, once nearly exterminated by overhunting, has responded to careful management and is now taken in fair numbers in the hardwood forests of the eastern United States. Grouse (including ptarmigan) are hunted throughout their range. (For further information on game birds, see *Micropædia*: HUNTING.)

NATURAL HISTORY

Habitat selection and food habits. As an order, the galliforms inhabit a wide variety of vegetational types, including dense and open forest, open grasslands, scrub and second-growth forest, and flooded riparian (river) forests. Megapodes live in dense jungle, some appearing in the open only to lay eggs on sandy beaches. The mallee fowl (*Leipoa ocellata*) is an exception, inhabiting the eucalyptus thickets that characterize the arid interior of Australia. The majority of galliforms roost on elevated perches at night, even those species that spend the daylight hours foraging on the ground. Virtually the only ones that live in treeless regions are certain of the grouse, such as the tundra-inhabiting ptarmigan, the prairie chickens (*Tympanuchus cupido* and *T. pallidicinctus*), the sage grouse

(*Centrocercus urophasianus*), and the sharp-tailed grouse (*Pedioecetes phasianellus*). The Eurasian black grouse (*Tetrao tetrix*) occurs in open country and in forested regions. Most of the true pheasant, including the peafowl, are residents of open forest with clearings. Guinea fowl of the genera *Guttera* and *Agelastes* and quail of the New World genus *Odontophorus* inhabit dense tropical forest. The hoatzin is closely associated with water and is almost never found far from it.

The food of galliforms is varied, most species being basically vegetarian, but they also take large numbers of insects, worms, and other invertebrates. Many use their feet to uncover food in leaf litter. The hoatzin feeds on leaves and fruit, especially those of arums (Araceae), rarely descending to the ground but occasionally entering the water to take small crabs or fish.

Polygamous and monogamous types

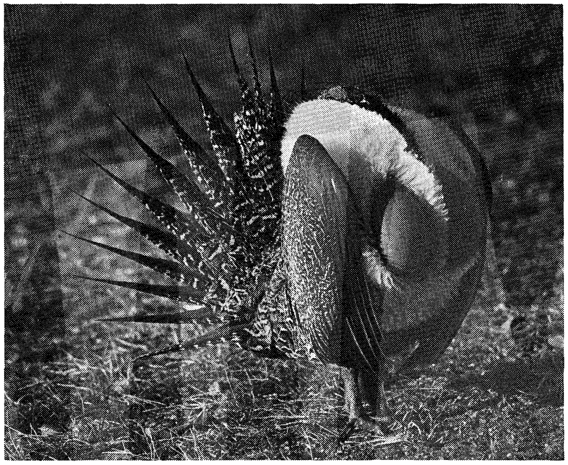
Courtship and mating. Gallinaceous birds vary considerably in reproductive behaviour, some exhibiting monogamous pair formation with a pair bond lasting through the breeding season, others showing varying types of polygamy, usually with members of both sexes being more or less promiscuous. Many, if not most, quail and partridge are monogamous, as are ptarmigan, guinea fowl, the hoatzin, some pheasant, and those megapodes and cracids that have been studied. Polygamy is known to occur in many grouse, including the North American grassland species, and in peafowl (*Pavo*) and some other phasianids. In the social displays of the grouse, a number of males assemble in a special assembly area, called a court, dancing ground, arena, or lek. The dancing ground lies outside of all nesting territories, and the same dancing ground is used year after year. Each of the males, which may number up to several dozen, has his own area within the chosen dancing ground. There he struts and postures, producing strange calls (some of which are produced in special esophageal air sacs) to attract the females, who visit the dancing ground to select a male and copulate. Dominant males occupy central positions in the arena group and may copulate with several females. Some species, such as the black grouse and the capercaillie (*Tetrao urogallus*), are variable in the degree to which males gather for display purposes, populations in more open areas tending more toward social displays. Males of the North American ruffed grouse (*Bonasa umbellus*) and those of some phasianids, the peacock pheasant (*Polyplectron*) and argus pheasant being examples, live in isolation, displaying to and copulating with any receptive female.

Nesting. The large majority of gallinaceous birds nest on the ground, the nest being only a shallow scrape, lined with soft grass or leaves. Cracids, the hoatzin, and the horned pheasant (*Tragopan*) build nests in trees, those of the cracids being relatively small for the size of the birds.

Construction of "incubators" by megapodes

The nesting of megapodes is unique among birds. The eggs of all species are placed in sand or soil, the heat for incubation coming from solar radiation, fermentation of plant matter, or even volcanic steam. Young megapodes, extremely precocious at hatching, dig their way to the surface and forage for themselves. They are fully feathered when hatched and can fly when 24 hours old. The simplest form of nesting, found in the maleo (*Macrocephalon maleo*), Wallace's megapode (*Eulipoa wallacei*), and some individuals of Freycinet's megapode (*Megapodius freycinet,* called jungle fowl in Australia), consists of placing each egg in a hole dug in sand to a depth of up to one metre (about one yard) in a site chosen for receiving the appropriate amount of solar radiation. Once laid and buried, the egg receives no further attention from the parents.

Harry Engels—The National Audubon Society Collection/Photo Researchers



Male sage grouse (*Centrocercus urophasianus*) in courtship display, revealing the bare air sacs in the sides of its neck.

Some members of the genus *Megapodius* (including jungle-dwelling members of *M. freycinet*) build mounds of decaying vegetation, up to 10 metres (35 feet) long and five metres high. Mounds built by the brush turkeys (*Alectura, Aepypodius,* and *Talegalla*) are smaller, up to about four metres in diameter and a metre high. The mallee fowl, the most studied of the megapodes, uses a combination of solar radiation and fermentation to maintain the incubation temperature.

In the mound-building species of megapodes, the male maintains the mound for much of the year. Prior to and during the incubation period, he opens the mound once or twice a day to control the generation, absorption, and radiation of heat. It is generally believed that the bird measures the temperature of the egg chamber when opening the mound. The temperature sensing organ is not known, but the most likely organ is the tongue. By instrumentation of mallee-fowl mounds, an Australian biologist, H.J. Frith, has found that the mound is maintained within a degree or two of 33° C (91° F) throughout the period of several months that there are eggs in it. Observations of one species of brush turkey, *Alectura lathami,* indicate that the frequent opening of the mound may be as important for ventilation as for temperature control.

Determinate versus indeterminate layers

The eggs of most gallinaceous birds are solidly coloured in white, buff, or olive, but those of species nesting in open areas are usually protectively coloured with blotches of brown or black. There is great variation in the number of eggs laid. Most members of the order are indeterminate layers, the female producing eggs until a certain number has accumulated in the nest. This characteristic has allowed man to exploit galliforms (especially the domestic hen, of course) for egg production. If the domestic hen were a determinate layer, as are members of many other bird orders, each individual would produce her clutch of about a dozen eggs and, regardless of whether or not the eggs were removed, would discontinue laying for at least two months.

The clutch size varies from two to about two dozen, the largest number occurring in megapodes. Partridge, quail, and the smaller pheasant lay from 12 to more than 20 eggs, but the larger pheasant, such as the crested argus (*Rheinardia ocellata*) and the great argus, normally lay only two eggs. The hoatzin and the cracids lay two or three eggs, three being the rule in the smaller cracids, such as chachalacas. Guinea fowl lay seven to 20 eggs, and turkeys eight to 18.

Incubation is usually performed by the female alone. Males of some New World quail, including the Montezuma quail (*Cyrtonyx montezumae*), the bobwhite (*Colinus virginianus*), and other members of the genus *Colinus,* incubate for short periods. The female of the southern European red partridge (*Alectoris rufa*) has been reported to lay two clutches, one incubated by the male, the other concurrently by herself. This behaviour, unique among birds, is in need of further investigation.

It is uncertain whether or not the male hoatzin incubates the eggs. Both sexes have been observed in nest building and taking care of the young, so it is possible that the male also assists in incubation.

Chick mortality

Care of the young. Young gallinaceous birds (except those of the hoatzin) are extremely precocious, walking and feeding within a few hours of hatching. Parental behaviour parallels mating behaviour: males of species in which a pair bond is formed usually assist in sheperding the young. Although clad in a protective coat of down and usually camouflaged with spots and streaks, the chicks suffer high mortality from predators and adverse weather.

Mortality rates of 50 percent or more are reported to occur in the period of a few months between hatching and independence of the young. Partridge, quail, and grouse maintain family groups (coveys) of a dozen or more birds that remain together until the next breeding season.

Hoatzin nestling wing claw

The hoatzin, unusual in many ways, also differs from other galliforms in that its nestlings are hatched practically naked and are fed by the parents in the nest. The young are fed regurgitated food from the crops of their parents. Although they have a long fledging period, young hoatzins scramble about in the branches around the nest, holding on with the clawed first and second digits of the wings and, in the manner of parrots, with their bills. The fact that the second digit of the wing (now considered to be digit III, in the evolutionary sense) is free and bears a functional claw was once thought (erroneously) to indicate affinities with the reptilian forebears of birds. Most authorities now believe that the free digits are a secondary adaptation that allows the young hoatzin great mobility during its flightless period and allows it to climb out of the water, into which it may deliberately drop when danger threatens.

Most gallinaceous birds reach sexual maturity at the age of at least one year. Some species, however, may be physiologically capable of reproduction at a much earlier age. The common quail (*Coturnix coturnix*), wild individuals of which normally breed at one year of age, matures to breeding condition in seven weeks in captivity. It is uncertain whether wild birds, hatched in the spring, actually do breed during the summer; environmental control factors, especially decreasing daylength, probably prevent the attainment of breeding condition by two-month-old birds in natural situations. Even the Congo peacock (*Afropavo congensis*), a relatively large species, is able to reproduce at one year of age when reared in captivity.

Vocalizations. Like many other birds that inhabit dense cover, forest galliformes are endowed with strong voices, ranging from musical whistles to harsh screams. The repertoire of most species includes alarm notes, food calls, "crowing" by males to advertise territories, and notes used to maintain social groups. Many species are especially vocal at daybreak; a few are reported to call at night. Megapodes give clucking and cackling calls during the day and mewing calls at night. Cracids produce loud calls that have immense carrying power. Males of some species have elongated, looped tracheae (windpipes), which are believed to add to the carrying power of the calls. Members of the social groups often cackle noisily together. Male peafowl utter a long mournful scream that sounds quite like a child in distress.

FORM AND FUNCTION

With the exception of the hoatzin, all galliforms have the same general body plan, being adapted for a primarily terrestrial existence. The feet and claws are large in all families, particularly so in the megapodes, reflecting their use for scratching and digging. The hind toe is larger and more functional in groups, such as the cracids, that spend much time in trees; it is smaller in the more terrestrial groups, but in none has it been lost, as it has in terrestrial birds of some other orders.

The short, rounded wings, powered by strong breast muscles (the white meat of the chicken), are indicative of the need for short, rapid bursts of flight, such as the escape from predators. Although no galliform is flightless, none is capable of long flights. The tail varies from extremely short (*e.g.,* in painted quail) to strikingly long; in many male pheasant the tail may be more than two-thirds of the bird's total length. The tails of some pheasant and of most megapodes are vaulted, having an inverted V-shape in cross section.

Ornamental plumage of the male

Male ornamental plumage is often remarkable in shape and coloration, combining spots or bars of silver, green, or purple iridescence with areas of brilliant orange, yellow, or white. Sculptured, fleshy wattles on the faces of male pheasant and grouse are often coloured a bright blue or red. In some male pheasant, such as the firebacks (*Lophura*), impeyans (*Lophophorus*), and peafowl, the head is ornamented with a small tuft of modified plumes, forming a tiny fan. Many cracids have patches of bare skin on the face or throat, usually red, yellow, or blue. Males of many curassows and guans possess head ornaments in the form of a brightly coloured fleshy knob or a bony casque (helmet) on the top of the head.

CLASSIFICATION

Distinguishing taxonomic features. The limits and interrelationships of galliform families have been determined on the basis of general body proportions, muscle and bone configurations, plumage, clutch size and egg characteristics, the appearance of the young, and some aspects of behaviour. Recent studies have utilized the biochemistry of egg white and blood proteins to indicate relationships.

Annotated classification. The classification presented below was established by the American ornithologist Alexander Wetmore, based on studies by many earlier workers.

ORDER GALLIFORMES

Suborder Galli

Family Megapodiidae (megapodes or mound builders)

All use heat other than body heat to incubate their eggs. Large feet, small heads. Tail often vaulted. Medium to large birds, 25–65 cm (10–25 in.); sexes alike. One fossil, *Chosornis praeteritus,* from the upper Pleistocene; 7 Recent genera, 12 species; Australia and East Indies to central Polynesia.

Family Cracidae (chachalacas, guans, and curassows)

Tail moderately long and broad. Plumage black or brown, duller in female. Most species with bare skin between eyes and beak (lores), some with fleshy wattles or other ornaments on face or crown. Medium to large; length 52–99 cm (20–39 in.); lower Eocene to present; 11 fossil and 11 Recent genera, 34 Recent species; New World tropics and subtropics, from southern Texas to Paraguay.

Family Tetraonidae (grouse)

Distinguished by having lower leg (tarsus) and sometimes feet at least partially feathered; nostrils also feathered. Tail strong; ornamental in some species. Many with brightly coloured bare skin over eye. Medium to large; 30–90 cm (12–36 in.). Lower Miocene to present; 3 fossil and 9 Recent genera, 17 Recent species; North America and northern Eurasia.

Family Phasianidae (pheasant, quail, partridge, and relatives)

Nostrils, feet, and (usually) tarsus unfeathered; esophageal air sacs lacking. Many species with spurs on the back of the tarsus. Plumage, especially of males, bright in many species, often with ornamental feathers; sexes usually different. Small to large; length 13–200 cm (5–80 in.); lower Miocene to present; 11 fossil and about 40 Recent genera, about 178 Recent species; virtually worldwide, except southern third of South America, northern Eurasia, and some oceanic islands.

Family Numididae (guinea fowl)

Distinguished by presence of small whitish spots on dark bluish or slate body feathers; head and neck bare or slightly feathered, often brightly coloured. Legs and feet large; spurs present only in *Phasidus* and *Agelastes.* Medium to large; length 43–75 cm (17–30 in.). No fossil species; 5 genera, 7 species; Africa south of the Sahara and Madagascar.

Family Meleagrididae (turkeys)

Distinguished by metallic sheen on feathers, bare head with red and blue skin and wattles or caruncles. Tarsal spurs of male long, of female short. Size large, length 84–110 cm (33–43 in.). Upper Pliocene to present; 1 fossil genus (Pleistocene) and 2 Recent genera, each with 1 Recent species; eastern and southern United States, Mexico, northern Guatemala, British Honduras.

Suborder Opisthocomi

Family Opisthocomidae (hoatzin)

Long neck and small head, with prominent crest of long feathers; beak short, stout, laterally flattened. Wings large and rounded; tail long and broad. Brown above, marked with white and russet; buff below. Nestling with 2 free, clawed digits on each wing. Adult size large, length about 61 cm (24 in.). One fossil genus (*Hoazinoides,* middle Miocene); 1 Recent genus with 1 species; river systems of northern South America.

Critical appraisal. Although the classification presented here has broad acceptance, modern authorities disagree on the numbers of species and, especially, genera in many groups. Forms once regarded as distinct are now often considered to be geographical races of larger species. The problem is particularly acute where island populations are involved, for the usual biological criterion of free interbreeding cannot be applied, and the taxonomist must infer from external appearance whether or not two forms belong to the same species. The genus *Megapodius,* for example, was considered by the American ornithologist James L. Peters, in his 1934 volume of the classic *Check-list of Birds of the World,* to comprise nine species with 25 subspecies. Another authority, Dean Amadon, in a study published in 1952, recognized only three species, with 15 subspecies.

At the generic level, taxonomy is more subjective than at the species level, the boundaries of the genus being subject only to the general views of the taxonomist as to what constitutes a genus. Recent years have seen a drift toward "lumping," with the result that what were once regarded as discrete genera are now frequently merged (lumped), a family often having fewer genera but about the same number of species. Details of the anatomy, behaviour, and distribution of many galliform species remain unknown. As more information about galliforms accumulates, further changes in the lower levels of classification may be expected. (F.H.)

Gruiformes (cranes, rails, coots, cariamas, bustards)

The bird order Gruiformes comprises a rather loose assemblage of 12 families that are generally agreed to be related but that differ widely in many aspects. They are an ancient group with a rich fossil history, but many are now restricted in range and few in number. Members of the order occur on every continent, but the only family with worldwide distribution is the Rallidae (rails, gallinules, and coots), with 132 living species. Cranes (Gruidae) are found on every continent except South America, but many of the 14 species have small populations, some on the verge of extinction. The bustards (Otididae), with 23 species, have a wide distribution, limited to the Old World, but hunting pressures and modern agricultural methods have greatly reduced their numbers. The mesites (Mesitornithidae), however, are confined to Madagascar, and the kagu (*Rhynochetus jubatus*) to the island of New Caledonia. Other small families in the order contain the hemipodes, or button quails (Turnicidae), plains-wanderer (Pedionomidae), limpkin (Aramidae), trumpeters (Psophiidae), finfoots (Heliornithidae), sun bittern (Eurypygidae), and cariamas (Cariamidae).

Although man's impact on them is very great, gruiform birds, because of their scarcity, have a negligible impact on man. With the possible exceptions of the sandhill crane (*Grus canadensis*), which descends on the grainfields of the Canadian prairies during the autumn migration, causing some crop damage, and the brolga, or Australian crane (*Grus rubicundus*), which causes similar damage in Queensland, no gruiform can be considered harmful to man's interests; some of the larger species, in fact, are hunted for food or sport.

GENERAL FEATURES

Size and weight

Gruiform birds range in size from the tiny button quails (*Turnix*) and miniature rails, such as the North American black rail (*Laterallus jamaicensis*) barely 15 centimetres (six inches) long, to the stately sarus crane (*Grus antigone*) of India, standing nearly 1.6 metres (five feet) high. The enormous kori bustard (*Otis kori*) and the Eurasian great bustard (*O. tarda*) may weigh up to 18 kilograms (about 40 pounds) and are the heaviest modern flying birds. Gruiforms vary widely in structure; some are adapted for life in or near water, others are adapted for life on land. Some forms fly well; but a number of gruiform species are flightless.

As a group, the gruiforms are probably best known for their impressive and graceful courtship displays, the most famous of which, the dances of the cranes, are imitated and adapted by many native peoples. The Ainu of Japan have a crane dance in honour of the Japanese crane (*Grus japonensis*), and many African peoples imitate the dance of the crowned crane (*Balearica pavonina*). Less well known, but no less spectacular, are the striking wing display of the sun bittern (*Eurypyga helias*) and the strutting and booming of the larger bustards.

NATURAL HISTORY

Ecology. *Habitat.* Gruiform birds live in a variety of habitats, from water and marshes to arid plains. The most aquatic are the finfoots and coots (*Fulica*). The former live along slow-flowing streams where heavy overhanging vegetation affords them cover, the latter on more open water. Most rails live in saltwater or freshwater marshes. The limpkin (*Aramus guarauna*) is essentially a marsh bird; in Florida it inhabits the sawgrass marshes and cypress swamps of the Everglades. Cranes bridge the gap between marsh and dryland birds, nesting in marshes but occurring in open plains and cultivated fields on migration and in winter. The sun bittern prefers muddy, wooded riverbanks but also occurs in woods well away from water. Trumpeters, mesites, the kagu, and some rails live in forest and dense brush. The remaining gruiform families inhabit more open country: the cariamas of South America favour grassland or hot, dusty plains with scattered bushes; the Old World bustards, button quails, and the plains wanderer prefer open, grassy plains, although they will accept old pastures and cultivated fields.

Diversity of diet

Food habits. Corresponding to the wide variety of habitats utilized by the gruiforms is the great diversity of food taken by them. In general, more animal than vegetable food is taken, but as a group the gruiforms are omnivorous. Coots and gallinules consume much aquatic vegetation. Finfoots live largely on mollusks, frogs, and small fish. The limpkin has a more specialized diet, consisting chiefly of certain large snails, which are broken open at certain favourite feeding stations that contain telltale piles of broken shells. The limpkins' near relatives, the cranes, eat almost any animal food, including rats, mice, moles, lemmings, lizards, snakes, frogs, tadpoles, snails, and a variety of insects. On its wintering grounds in Texas, the whooping crane (*Grus americana*), an endangered species, lives largely on crustaceans. The sandhill crane includes berries and grass in its otherwise animal diet on the northern tundra and gleans some plant material from old potato and grain fields on its southern wintering grounds. The Asiatic cranes that winter in Japan (such as the hooded crane, *Grus monacha,* and the white-naped crane, *G. vipio*) glean grain in rice paddies, and Japanese cranes in Hokkaido are fed corn (maize) by the local farmers. Cranes use their powerful bills for digging in the ground to get at bulbs and roots below the surface, and a similar foraging action has been noted for the gray-necked wood rail (*Aramides cajanea*) of tropical America, which probes in ground debris, flicking it aside with the bill. Rails as a group, like cranes, are omnivorous, though the bulk of their diet consists of small marsh animals, such as snails, crustaceans, frogs, and water insects. The purple swamphen (*Porphyrio porphyrio*), a vegetarian, often has a feeding platform on which it stands and pulls up the surrounding water plants; it will also climb up reeds and eat the flower heads.

The other members of the order have a similar mixed animal and vegetable diet, though with the emphasis always on the animal side. Captive cariamas readily eat meat, and a captive kagu feasted on ground steak provided by its owner. Bustards are fond of grasshoppers, and their varied diet also includes dung beetles, termites, centipedes, grass, clover, vegetable crops, and even, in Africa, the gum from the trunks of *Acacia* trees. The sun bittern stalks insects as a heron stalks fish, stealthily approaching its prey with neck drawn in, then grabbing it with a sudden, stabbing thrust of the bill.

Dancing by cranes

Reproductive behaviour. *Courtship.* The breeding cycle of many gruiform birds begins with elaborate courtship rituals and displays. Cranes pair for life, and the strong pair bond necessary to maintain this partnership is initiated and continued by a series of displays that, since they often consist of two birds facing each other and leaping into the air, are generally known as dances. The ceremony frequently begins as two birds circle each other with a curious, formal step, the legs stiff and the head and neck held high. The next action is bowing or "head bobbing," in which the head is held horizontally, the neck curved down in a U, and in this position the head and neck are bobbed or pumped up and down. One bird may do

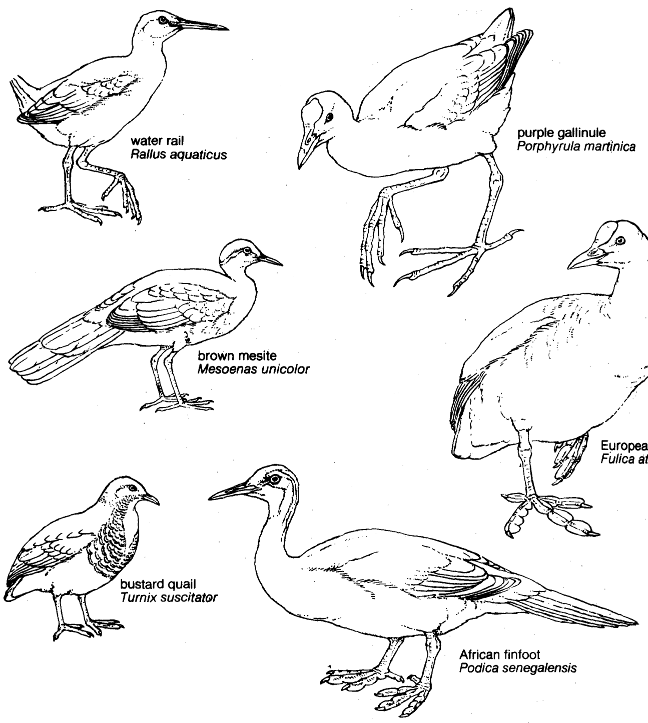


Figure 26: Body plans of some smaller gruiforms.
Drawing by R. Keane

this while the other bird looks on, or both may do it together. One bird (or both) then turns in a circle on the spot while continuing to bow. The momentum increases until suddenly both birds leap into the air and follow this by dancing. While dancing, the birds often pick up some object from the ground and toss it in the air as they leap. The object is usually something used in nest building, such as a stick, leaf, or tuft of grass, and the action may be related to nest-building drives. Again, while leaping in the air one bird may turn its back on the other. Another beautiful display involving a pair of cranes is the "duet." The birds first circle each other with the same formal step that initiates the dance, but instead of going into the dance they droop their wings, throw back their heads, and call in unison.

Although dancing intensifies at the beginning of the breeding season and is doubtless primarily connected with courtship and pair formation, it can occur at any time of year and may have other functions as well. Dancing often seems to be a method of releasing pent-up energy, as when a bird dances on its own or when a dancing pair sets off dancing in nearby individuals of a flock. So strong is the instinctive urge to dance that a five-day-old chick is recorded as leaping up and down and going through other motions of the dance, even though it had never seen another crane.

Vocalizations of rails

Pair formation in many rails is effected chiefly by voice; males establish territories and defend them vigorously with distinctive songs and calls while at the same time attracting females. Rails produce an amazing variety of noises. In North America, the clapper rail (*Rallus longirostris*) has a loud cackling call, the sora (*Porzana carolina*) an explosive whinny, and the yellow rail (*Coturnicops noveboracensis*) makes a noise like two small stones being clicked together. In Europe, the water rail (*Rallus aquaticus*) squeals like a stuck pig, and the corncrake (*Crex crex*) produces a rasp like a heavy comb being drawn over a piece of wood. The gray-necked wood rail greets the dawn with a ringing "Pop-tilly, pop-tilly, ko-ko-ro-ko-ko," which has given it the local name of "cocaleca" in Panama. In Africa, a large rail, *Himantornis haematopus,* of the Congo forests makes a pumping sound, and the black crake (*Limnocorax flavirostra*) in the papyrus swamps makes a curious gurgling sound. The tiny crakes in the genus *Sarothrura* have a variety of melodious calls, the buff-spotted crake (*S. elegans*) making a low moaning noise like the sound of a tuning fork. More familiar are the plebeian squawks and grunts of the wide-ranging coots and gallinules.

Rails also have courtship displays involving lowering the head, raising the wings, and fanning the tail, often while uttering some special call. Coots and gallinules display their colourful frontal shields and fan their tails to show off the white under tail coverts. Other species circle their partners in various postures. The display often ends with the male chasing the female, and the chase may end in copulation. In button quails the role of the sexes is reversed. The female is the most brightly coloured and courts the male, erecting her tail, puffing out her neck, and running around him while uttering a low crooning note. The male sun bittern selects an open spot, often in a patch of sunlight, where he spreads and raises his tail and wings until they meet in front of his head, exposing elegant patterns of red-brown, olive, gray, and black. In this posture he runs in a circle and may jump up in the air or bob his head.

In its aggressive display, the kagu stands erect, with the long feathers of the crest raised, the wings held out from the body, and the tail drooped. In this posture, it bounces at the adversary. In a playful mood the kagu will toss sticks and stones around with its bill, in a manner reminiscent of cranes. A captive cariama has been seen to run at a tree with lowered head, jerking its tail and giving a short cluck, then striking the tree with both feet. In spring wild birds have a bustard-like display (see below) accompanied by loud yelping calls.

Vocal displays

Trumpeters, named for their loud, resonant cries, have a cranelike dance involving strutting around on the ground and leaping into the air. Males of the larger bustards, the

Drawing by R. Keane

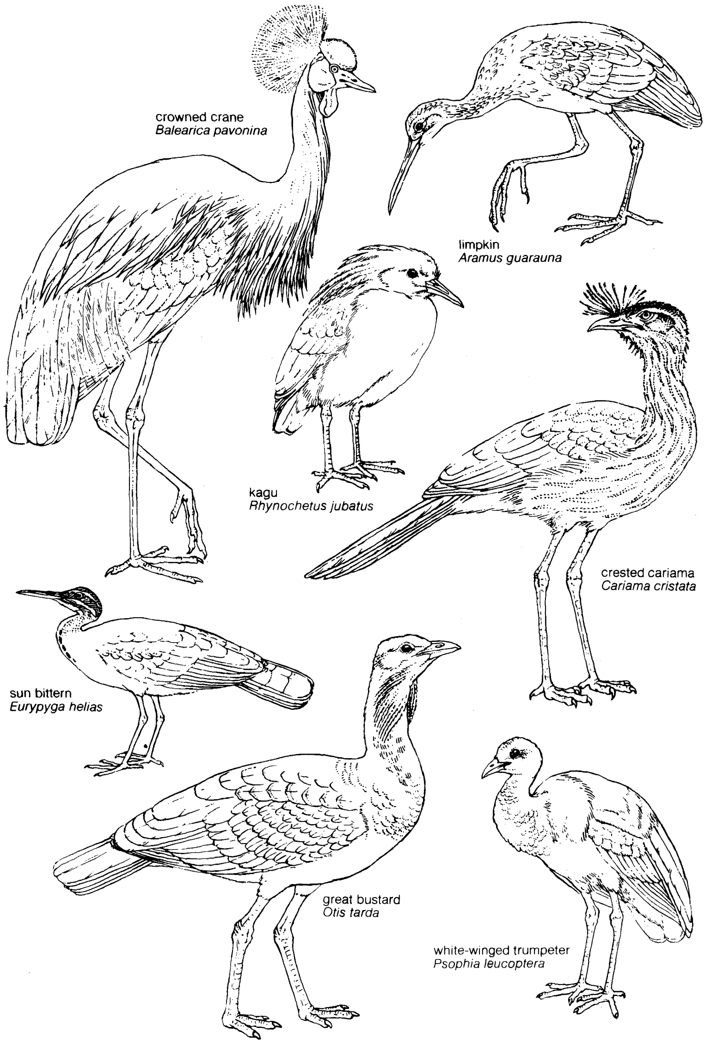
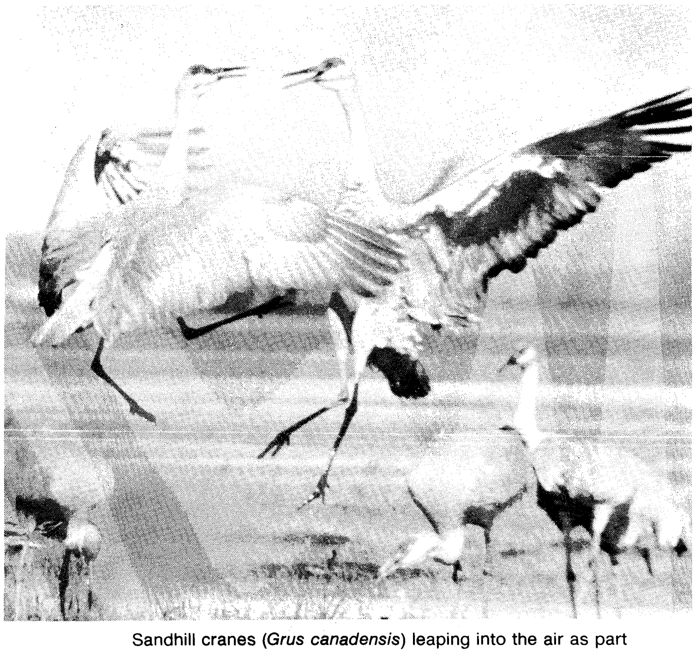


Figure 27: Body plans of some larger gruiforms.

great and kori bustards, for example, develop a special pouch in the neck during the breeding season with which they produce loud, booming calls. The Australian bustard (*Ardeotis australis*) "booms" with closed beak, producing a sound somewhat like the distant roar of a lion. Special display feathers that grow on the head, neck, and breast are molted after the breeding season. A large bustard in full breeding regalia, with all feathers puffed out, is scarcely recognizable; the head is immersed in an immense ruff, and the wings and tail are raised until they meet over the back, the total effect being that of an enormous feather ball. The ball then struts around in front of a female, who feigns indifference. Smaller bustards have somewhat different displays. The crested bustard (*Lophotis ruficrista*) of Africa has an aerial display flight in which it rises about 100 feet into the air and then planes steeply back to earth.

Nesting. Gruiform birds nest both on the ground and in bushes and trees. Of the ground nesters, button-quails and the plains wanderer nest in a grass-lined hollow, often building a domed roof and side entrance. Cranes raise up a pile of vegetation in open shallow water, and the limpkin builds a concealed nest in dense marsh grasses. Marsh-dwelling rails build simple nests of grass and aquatic plants, often in a thick tuft of grass, the blades of which they pull down over the nest to conceal it. The kagu builds a nest of sticks and leaves in a depression in the ground. Bustards scarcely have a nest at all, the eggs being laid on bare ground, sometimes beneath a bush or clump of grass for concealment. Finfoots and the sun bittern prefer a nest site on a branch of a tree, and finfoot nests generally overhang water. Of the two species of cariamas, one nests on the ground and the other in bushes. Trumpeters are variously reported as nesting on the ground and in holes in trees. Mesites place their simple stick nests a metre or two up in a tree. Being flightless, they must always find a site where a connecting series of branches leads from the ground to the nest, enabling them to hop up to it.

By courtesy of the U.S. Bureau of Sport, Fisheries, and Wildlife; photograph, David Marshall



Sandhill cranes (*Grus canadensis*) leaping into the air as part of the courtship dance characteristic of this family of birds.

Incubation of the eggs

Clutch sizes vary widely within the order, from the single egg of the kagu to over 10 in some rails. Exceptional clutches of 15–20, recorded for some coots and gallinules, may be due to more than one female laying in the same nest. The eggs are usually white or buff, sometimes pale gray or pale green, immaculate in a few species, but usually with brown spots or blotches at one end. The incubation period is typically about three weeks, extending to four weeks in the cranes and larger bustards. Both sexes incubate the eggs and care for the young when hatched in all families except bustards, trumpeters, and button quails. Button quails are polyandrous (an individual mating with several members of the opposite sex), and the tasks of incubation and care of the young are performed entirely by the males. In bustards and trumpeters, the female does all of the incubating and caring for the young. Young gruiforms are downy, plain black in rails and dark brown in limpkins, variously patterned in most other groups. They leave the nest immediately or very soon after hatching, except for those of the sun bittern, cariamas, and bustards, which are cared for at the nest for a short time. There is some evidence that in the trumpeters, which habitually travel in flocks in the adult state, several pairs may pool their young and look after them communally.

Molt. The sequence of molts and plumages is very poorly known except in cranes, the young of which have a brown or gray juvenile plumage, with white-tipped or blackish feathers in some species. The juvenile feathers are gradually replaced at each successive molt with the white or gray feathers of the adult, but the last browntipped feathers do not completely disappear until about the third summer, when the bird is a little over two years old. Adult cranes, at least those nesting in northern temperate zones, molt in two stages; many of the wing and tail feathers are molted in early summer, at which time the birds may be flightless for a while; the rest of the plumage is gradually molted between August and October.

Migration and locomotion. Most gruiforms are nonmigratory. Bustards, button quails, and the plains wanderer migrate locally, following the rains to feed and nest. Only the birds nesting in the North Temperate Zone are true migrants; this group includes many cranes, some rails, and the Eurasian bustards. The spectacular migrations of cranes have excited man's interest since earliest times. The peoples of eastern Asia welcome the return of the cranes as symbolic of the coming spring, and in fall the farmers of Japan welcome the birds back to the rice paddies where they spend the winter. In Japan, the cranes return every year to the same traditional wintering grounds, where they are given strict protection. The last remaining whooping cranes in North America are likewise carefully looked after and are counted at the Aransas National Wildlife Refuge in Texas on their return from the north. Most cranes cover great distances on migration. Sandhill cranes travel 4,000 miles from their nesting areas in Alaska and eastern Siberia to southern United States, and common cranes (*Grus grus*) and demoiselle cranes (*Anthropoides virgo*) cover similar distances in the Old World.

When the northern marshes freeze over in winter, rails are forced to head south. Amazingly, these birds that fly so weakly across a marsh, with floppy wings and dangling legs, are able to travel thousands of miles each year on migration. The corncrake, nesting in Scotland, may winter in South Africa, and the sora rail of North America regularly crosses 800 miles of ocean to reach Bermuda.

The Eurasian bustards travel shorter distances, going only far enough to escape bitterly cold weather. Some species form large flocks at migration time. Bustards are great walkers, and local migrations in Africa are for the most part performed slowly on foot.

Travel on foot

Gruiform birds have a predilection for travel on foot. Many only fly when pressed, and some, like the mesites, have become flightless or nearly so. Many oceanic islands have been colonized by rails, which then evolved flightlessness in the absence of predators. The subsequent advent of rats, cats, pigs, or goats on such islands, usually with accidental or intentional assistance from man, has resulted in the extinction of a number of such rails. Rails typically sneak away on foot in thick vegetation, and button quails are equally loath to fly, preferring to walk away with their quick, nervous gait, stopping every so often to raise their heads and look around for danger. Trumpeters run fast and can even swim. Finfoots spend most of their life in water and prefer to hide in thick riverside bushes when disturbed, rather than fly. On the ground, gruiform birds move efficiently and even elegantly. The sun bittern walks gracefully with slow, precise steps, its neck outstretched. Cariamas run swiftly over the plains. Rails have a very characteristic walk in which the tail is flicked up with each step, and both the limpkin and the kagu share this tail-flicking action. Even such fine fliers as the cranes prefer walking, and there is no more elegant sight in the avian world than a tall and stately crane walking with deliberate and dignified gait across the prairie. In flight, cranes and the limpkin have a charcteristic wing action—a slow downstroke followed by a quick, flicking upstroke.

FORM AND FUNCTION

Gruiform birds vary greatly in shape and size and exhibit a broad range of morphological characteristics. Their plumage is predominantly brown or gray. Some have brightly coloured soft parts such as the bare, red skin on the head and neck of some cranes, used in displays, and the bright red and yellow bills and frontal (forehead) shields of gallinules. The crowned crane has a curious crest of stiff, golden feathers. The sexes are alike in most groups, except among the button quails, in which the female is more brightly coloured, and the bustards, where the males are more colourful.

The wings are rounded and often long, although in the nearly flightless mesites they are greatly reduced. The length of the tail varies, being proportionately short in button quails, rails, and trumpeters and rather long in mesites, finfoots, and the sun bittern. Cranes have very long inner secondary feathers (those of the inner wing or "forearm"), which extend beyond the end of the tail, giving the impression of a long-tailed bird. The bill is generally long and slender, particularly so in cranes, many rails, and the limpkin, although the cariamas have hooked bills which are doubtless used in tearing up mammalian prey. The legs are rather long, reflecting a preference for walking. The toes vary greatly—in the finfoots and coots they are lobed for swimming, in rails and the limpkin they are long and slender for walking on lily pads and other aquatic vegetation, but bustards and cariamas have short toes for running on hard surfaces. The hind toe, when present, is usually elevated.

Special anatomical features

Some groups have anatomical features peculiar to themselves. The mesites possess five pairs of powder down patches, far more than any other group, and the function of these is uncertain. Cranes, and the adult male limpkin have an extremely long trachea, or windpipe, that is coiled in several convolutions. These convolutions of the trachea probably give added power and resonance to the voice, which can carry for distances of a mile or more. Rails have a laterally compressed body, which gives rise to the expression "thin as a rail," enabling them to sneak between reeds and blades of grass without telltale movements of the vegetation. Most young rails have a claw at the tip of the alula (the "bastard wing" or "thumb") that enables them to clamber around on marsh vegetation. Finfoots have a sharp spur of uncertain function at the bend of the wing. The kagu, like the mesites and some rails, is flightless, a condition that may lead to its extinction by the dogs, cats, pigs, and rats that have been introduced on New Caledonia. Males of great and kori bustards have a gular (throat) pouch during the breeding season that opens into the mouth under the tongue and can be inflated at will. It is used by the birds to produce booming calls during courtship. The Australian bustard has no gular pouch, producing its calls by filling the esophagus with air. The esophagus is similarly used in sound production by the button quails and by rails of the genus *Sarothrura.*

EVOLUTION AND PALEONTOLOGY

Gruiform birds have the best fossil record of any avian order, going back to the Late Cretaceous, nearly 100,000,000 years ago. Fourteen fossil families are known, divided into 86 genera and 146 fossil species. In addition, 39 extant species are also known as fossils.

The oldest gruiform bird is the crane-sized *Laornis* from the Upper Cretaceous of New Jersey, recently shown to be closely related to the rails. *Telmatornis,* from the same epoch, formerly considered a gruiform, is now placed closer to the thick-knees (Burhinidae) in the order Charadriiformes.

Next in age are two families from the Paleocene (about 60,000,000 years ago), the Gastornithidae and the Diatrymidae, the latter containing some fascinating giant species—*Diatryma steini* stood about seven feet tall and had a massive head and bill—but it is in the Eocene, starting about 54,000,000 years ago, that gruiform birds first became abundant and the first representatives of modern families appeared: rails, cranes, and a bustard. In the Oligocene the limpkins and the suborder Cariamae had their beginnings. The Cariamae are represented today by only two living species, but their fossil history shows that in earlier epochs they were a more widespread and successful group. They include a number of flightless giants, the best known of which are several species of *Phororhacos* from the lower Miocene of Patagonia. These were powerful birds, the largest in excess of 2.2 metres (seven feet) in height. The best known form, *P. longissimus,* must have been a formidable predator, having a massive skull 65 centimetres (two feet) long and 25 centimetres (10 inches) high, with a hook at the tip of the beak.

With the exception of the above, and a single *Turnix* from the Pleistocene of China, the smaller modern gruiform families have no fossil records. Nevertheless, the abundance of fossils shows what an amazingly successful group the gruiforms were. They are known from all six continents, and for every fossil species discovered there must have been tens or even hundreds that are unknown. Today, however, gruiform birds are on the decline. Of the 12 gruiform families, fully eight are represented by three species or less, and four have only a single species. The only one that is at all numerous is that of the rails, with about 130 species. Many gruiform birds appear in the "Red Data Book" of endangered species, published by the International Union for the Conservation of Nature. Five of the 14 extant cranes are listed as in danger, the best known being the whooping crane. Seven rails are listed as in danger and 15 more have become extinct since about 1600. Many of the extinct rails lived on small islands and a number had become flightless, making them vulnerable to man and his animals. Nevertheless, the Gruiformes had been declining long before man entered the scene, apparently because the ecological conditions that favoured them in the past are less available to them today.

Endangered species

CLASSIFICATION

Distinguishing taxonomic features. The order Gruiformes is a heterogeneous group bound by taxonomic characters that are largely anatomical, and hence not readily evident on the live bird. Few are common to all members of the group, and no single character serves to separate the Gruiformes from all other orders. Some of the features that have been used in classifying the order are: the condition of the feet and toes; the type of palate, osteologically; the type of pelvic musculature; the number of carotid arteries; the presence (eutaxy) or absence (diastataxy) of the fifth secondary flight feather in the wing; the number of wing and tail feathers; the number of cervical vertebrae; the type of nostrils; the presence or absence of an aftershaft (a small "second feather" attached to the shaft of a body feather); and the behaviour of the young.

Annotated classification. The classification presented here is based on work by the ornithologists Alexander Wetmore, Pierce Brodkorb, and Joel Cracraft. Groups indicated by a dagger (†) are known only as fossils.

ORDER GRUIFORMES

Primarily marsh-dwelling birds of medium to large size. Toes not webbed (lobed in a few genera). Hallux (hind toe) usually elevated, sometimes absent. Two carotid arteries usually present. Aftershaft usually present. Crop lacking. Young usually nidifugous (precocious).

Suborder Mesitornithides

Distinguished from other suborders by reduced clavicles (collarbones) and presence of five powder-down patches.

Family Mesitornithidae (mesites). No fossil record. Superficially dovelike, but with characters of order and suborder. Hallux well-developed and functional. Three species; terrestrial, in forest and dry brush; confined to Madagascar; length 25–27 cm.

Suborder Turnices

Principal difference from other suborders is the possession of a well-developed basipterygoid process (a projection at the base of the skull).

Family Turnicidae (button quails). Pleistocene (one specimen) to present. Small and quail-like; short legs, no hallux;

beak short, slightly downcurved. Unique in the order in having only one carotid artery. Female larger and more brightly coloured than male. Eggs roundly oval. Eutaxic. Fifteen species; in grasslands and brush; southern Europe, Asia, Africa, and Australia; body length 11–19 cm.

Family Pedionomidae (plains wanderer). No fossil record. Similar to button quails, but differs in the following details: hallux present; two carotid arteries; wing diastataxic; eggs pointed. One species, confined to dry plains of Australia; body length about 16 cm.

†Suborder Gastornithes
Fossil only.

†*Family Gastornithidae.* Upper Paleocene of Europe; five species; large, flightless.

†*Family Diatrymidae.* Upper Paleocene to middle Eocene; North America and Europe; four species; large flightless predators.

Suborder Grues
Adults with down feathers in both pterylae (feather tracts) and apteria (areas between tracts); hallux present and variable; oil gland present and tufted.

Superfamily Gruoidea (cranes, limpkin, and trumpeters)
Only gruiforms with supraorbital furrows (*i.e.,* above the eye); distinguished from others by pelvic muscle formula. Ten to twelve tail feathers.

†*Family Geranoididae.* Fossil only; lower to middle Eocene; North America; seven species known.

†*Family Bathornithidae.* Fossil only; lower Oligocene to lower Miocene; North America; eight species known.

†*Family Eogruidae.* Fossil only; upper Eocene to upper Miocene; eastern Asia; two species known.

Family Gruidae (cranes). Lower Eocene to present. Bill long and straight; neck and legs long; hallux small and elevated. Head usually partly naked in adult. Wing diastataxic. Caudofemoral muscle present (except in *Balearica*). Twenty-one fossil and 14 Recent species; plains and marshes of the world, except South America; length about 79 to 150 cm.

†*Family Ergilornithidae.* Fossil only; lower Oligocene to lower Pliocene; eastern Europe and Asia; three known species.

Family Aramidae (limpkins). Lower Oligocene to present. Bill long; neck long and slender; head feathered. Hallux large and functional. Caudofemoral muscle absent. Swamps and marshes of New World tropics and subtropics; five fossil and one Recent species; length 58 to 71 cm (in Recent forms).

Family Psophiidae (trumpeters). No fossil record. Bill short. Separated from Gruidae and Aramidae by nostril shape, eutaxic wing, absence of occipital foramina (perforations at the base of the skull). Three species, in lowland forest of South America; length 43 to 53 cm.

Superfamily Ralloidea
Lack occipital foramina and supraoccipital furrows. Hallux well-developed and functional; cervical vertebrae fewer than in Gruoidea.

Family Rallidae (rails, gallinules, and coots). Upper Cretaceous to present. About 56 fossil and 132 Recent species; principally in marshes; worldwide; length about 14 to 51 cm.

†*Family Idiornithidae.* Fossil only. Upper Eocene to lower Oligocene of Europe and North America; eight species known.

Suborder Heliornithes
Distinguished from all other gruiforms, except coots (*Fulica*), by possession of lobed feet; by pelvic muscle formula, type of flexor tendons, and possession of 18 tail feathers.

Family Heliornithidae (finfoots). No fossil record. Three species; slow-flowing streams in tropical areas of Central and South America, Africa, and India and Southeast Asia; length 30 to 62 cm.

Suborder Rhynocheti
Large aftershaft; powder downs occurring as scattered groups of feathers. Large operculum (covering flap) over nostrils. Young nidicolous (dependent).

Family Rhynochetidae (kagu). No fossil record. One species, confined to forested highlands of New Caledonia; length about 56 cm.

Suborder Eurypygae
One pair of powder-down patches. Oil gland nude. Eighteen cervical vertebrae. Young nidicolous.

Family Eurypygidae (sun bittern). Beak medium length, straight, sharp; neck slender. Tail of medium length. One species; margins of woodland streams in Central and South America; length about 46 cm.

Suborder Cariamae
Distinguishing features in palate structure; talon-like nail on second toe; and type of flexor tendons.

†*Family Cunampaiidae.* Fossil only. Lower Oligocene of Argentina; one known specimen.

†*Family Brontornithidae.* Fossil only. Lower Oligocene to middle Miocene of Argentina and Uruguay; two species known.

†*Family Palaeociconiidae.* Fossil only. Lower Oligocene to middle Pliocene of Argentina; four species known.

†*Family Prophororhacidae.* Fossil only. Lower Pliocene to lower Pleistocene of Argentina; three species known.

†*Family Phororhacidae.* (Phorusrhacidae of some authors). Fossil only. Lower Miocene to lower Pleistocene; 10 species, nine from Argentina, one from Florida. Medium to large flightless predators.

†*Family Psilopteridae.* Fossil only. Lower Oligocene to middle Pliocene of Argentina; nine species known.

Family Cariamidae (cariamas or seriemas). No fossil record. Moderate-sized cursorial birds; legs long, feet small; tail long; beak broad, moderately long, slightly decurved. Forehead and back of neck crested. Two species; in grassland and brush, respectively, of east-central South America; length 76 to 92 cm.

Suborder Otides
Oil gland absent. Scales on tarsus (lower leg) hexagonal; hallux absent. Sternum with two pairs of notches. Wing with 11 primary flight feathers. Egg-white protein structure (of Otididae) unlike those of other gruiforms.

Family Otididae (bustards). Cursorial, but strong flying birds of open plains of Eurasia, Africa and Australia; three fossil and 23 Recent species; length 37 to 132 cm.

†*Family Gryzajidae.* Fossil only. Lower Pliocene of the Ukranian S.S.R.; one species known.

Critical appraisal. The great diversity in the order Gruiformes is reflected in the uncertainty of taxonomists about gruiform relationships. The most widely accepted classification is that of the American paleontologist Alexander Wetmore, used above, with 12 Recent families in one order. Some authorities place the plains wanderer (*Pedionomus*) in Turnicidae. A German ornithologist, Erwin Stresemann, underlining the differences between the gruiform families, divided them into ten orders, most of them monotypic (*i.e.,* with only a single form). A British anatomist, P.R. Lowe, dismayed by the wide variety of characters in the Gruiformes, eliminated the order altogether, distributing the families among other orders, a decision later repeated by a French worker, R. Verheyen, who divided the families among five orders, some of which are not recognized today. Verheyen's order of Gruiformes contained only the Psophiidae, Aramidae, Gruidae, and Otididae. In an attempt to resolve this confusion, an American biochemist, H.T. Hendrickson, studied the egg-white proteins of 10 of the 12 gruiform families (he had no material for Pedionomidae or Mesitornithidae) and found that a close group was formed by five families—Eurypygidae, Heliornithidae, Rallidae, Turnicidae, and Psophiidae. The Aramidae bridge the gap between this group and the Gruidae, but the remaining families (Rhynochetidae, Cariamidae, and Otididae) seemed to be very different, the last so different that they may be of independent origin from the rest of the order. (G.S.Ke.)

Charadriiformes (plovers, sandpipers, gulls, terns, auks)

The order Charadriiformes comprises the shorebirds, gulls, auks, and their relatives. These birds form an important and familiar segment of the avifauna of the world's coasts and inland waterways, of the Arctic regions, and of the oceans and their islands. They are mostly strong-flying birds of open country or open water, nesting on the ground and feeding on animal food in or near water. The order is worldwide in distribution, and some species perform the most extensive migrations of any birds.

GENERAL FEATURES

The order is a heterogeneous assemblage of 16 families, linked by similarities in anatomical features (especially skeleton and plumage) and developmental pattern. Better known members of the order fall into three groups,

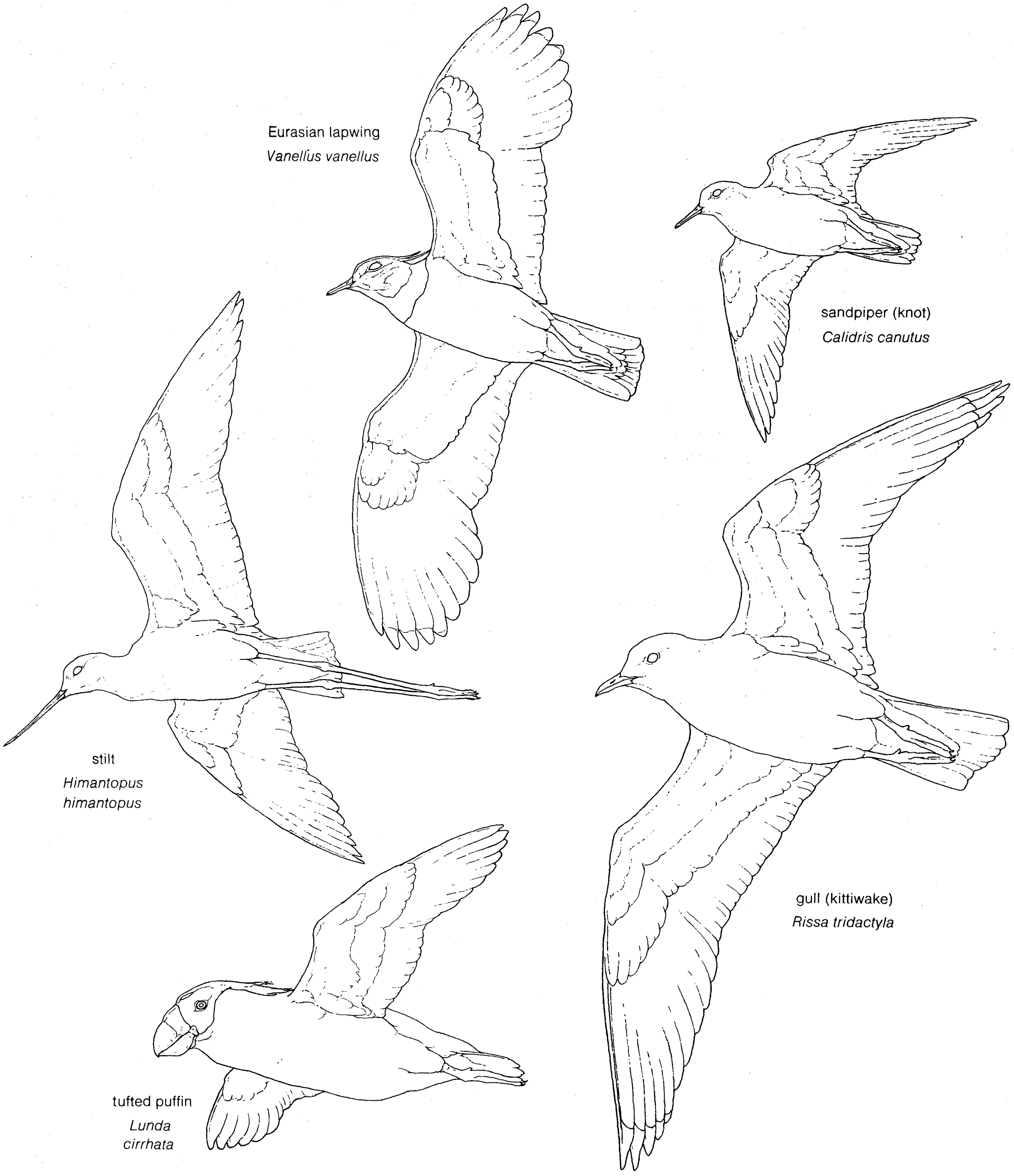


Figure 28: Body plans of typical members of major charadriiform groups.
Drawing by G. Tudor

easily recognized on the basis of general body plan. The first of these (the suborder Charadrii), collectively known as shorebirds or waders (Figure 28), includes sandpipers, plovers, lapwings, snipes, stilts, and some less familiar forms. They are primarily birds of shorelines and other open areas, and they walk or wade while feeding. There are about 200 species, varying in size from the least sandpiper, a sparrow-sized bird of about 20 grams (slightly under an ounce), to large curlews of about 640 grams (1½ pounds, near the body size of a small chicken). A second group, the suborder Lari, contains 92 species of gulls, terns, skimmers, skuas, and jaegers. They are long-winged, web-footed birds, the smallest of which is the least tern (*Sterna albifrons*), weighing about 43 grams (1.5 ounces), with a wingspread of about 50 centimetres (20 inches). The largest, the great black-backed gull (*Larus marinus*), weighs about 1,900 grams (a little over four pounds) and has a spread of about 165 centimetres (65 inches). The third and smallest suborder, Alcae, contains 21 species of auks, murres, guillemots, and puffins, all in a single family, Alcidae. They are compact, streamlined, marine birds with short, narrow wings and webbed feet. Alcids are adapted for swimming on the ocean surface and underwater.

Most charadriiforms have plumage patterns in white, grays, browns, and black, and many have bright red or yellow feet, bills, wattles, eyes, or mouth linings. A few species have both dark and light plumage phases.

Distribution

Each of the larger families (Laridae, Charadriidae, Scolopacidae) is practically worldwide in distribution, although none of the Scolopacidae breeds in Australia. The skuas and jaegers (Stercorariidae) are found in high latitudes of both hemispheres and are wide-ranging through the world's oceans. Auks and their allies (Alcidae) are widespread in the oceans, islands, and seacoasts of the Northern Hemisphere. They are not related to their similar counterparts of the Southern Hemisphere, the diving petrels and penguins. The oystercatchers (Haematopodidae) are found on coasts of all continents except Antarctica and occur inland in Europe and Asia. A group of families occurs in tropical (or tropical and temperate) regions of the Eastern and Western hemispheres: jacanas (Jacanidae), painted snipes (Rostratulidae), avocets and stilts (Recurvirostridae), thick-knees (Burhinidae), and skimmers (Rynchopidae). The coursers and pratincoles (Glareolidae) occur throughout tropical and temperate regions of the Old World, and the crab plovers (Dromadidae) are limited to shores of the Indian Ocean. Seed snipe (Thinocoridae) are found in southern South America, and northward in the Andes; sheathbills (Chionididae) occur on islands of the southern Atlantic and western Indian oceans and on the southern coast of South America and adjacent Antarctica. Phalaropes (Phalaropodidae) breed in northern regions, and two species winter at sea.

Importance to man. The eggs of murres, puffins, gulls, terns, and lapwings have long been harvested for food. These birds are particularly suitable for such use because many nest in enormous colonies, and because they replace the clutch if the first is taken soon after laying. Several hundred thousand eggs, and sometimes over a million, may be taken from a locality in a single year. Certain colonies, especially those of gulls and terns, have been raided without regard to the future of the colonies, but carefully controlled egging has long been conducted in the Faeroe Islands, Iceland, Greenland, and other northern regions. Adult puffins and other alcids are also harvested with long-handled nets on the Faeroes and in Iceland.

Extravagant exploitation of the great auk for food on its North Atlantic nesting islands by sailors, and later slaughter for the feather trade, probably caused its extinction in the 1840s. Other charadriiform birds, especially terns and gulls, assumed a sudden economic value for decorating women's hats in the latter half of the 19th century, and raiding of breeding colonies in North America almost extirpated several species. Aroused public opinion, hastened by the activities of the newly formed Audubon societies, brought protection to gulls, terns, and other species.

Charadriiform birds have had considerable economic impact in various other ways. California gulls are credited with saving the pioneers' crops in Utah during a plague of locusts, and today gulls habitually follow the farmer's plow, consuming exposed grubs and mice. Flocks of noddy terns and other birds serve to guide Hawaiian fishermen to schools of tuna, and the numbers, kinds, and behaviour of the birds may also indicate the size of the fish and the size and depth of the school. Certain shorebirds were once extensively killed for food or sport (causing near extinction of the Eskimo curlew); today woodcock and snipe are hunted under regulation.

Gulls and shorebirds are occasional hazards at airports, where airplanes have been damaged by collisions.

The order as a whole has been the subject of much scientific investigation, leading to important studies on speciation, ecology, ethology, migration, anatomy, and physiology.

NATURAL HISTORY

Locomotion and feeding behaviour. Most shorebirds inhabit open areas and are strong fliers, some performing extensive migrations that cover long distances over water. A ruddy turnstone (*Arenaria interpres*) banded in the Pribilof Islands was recaptured in the Hawaiian Leeward Islands, 3,770 kilometres (2,325 miles) away, four days later. Gulls, terns, skimmers, skuas, and jaegers spend much of their time on the wing, both in migration and in moving within the breeding or wintering grounds. Immature sooty terns (*Sterna fuscata*) spend several years flying at sea before first coming to land to breed, and Arctic terns (*Sterna paradisaea*) fly each year to and from Antarctic waters after breeding in the Arctic. Gulls are given to soaring and gliding more than the others and are the only members of the suborder Lari that spend considerable time swimming or resting on water. Of the three major charadriiform groups, the alcids spend the least time on the wing, but they are strong, fast fliers for short distances. Outside the breeding season they are pelagic (*i.e.,* living on the open ocean).

Food habits of shorebirds

The feeding habits of the charadriiforms are as varied as their external appearance. Jacanas inhabit pools and lakes thick with water lilies and other aquatic vegetation. They run agilely on lily pads with prancing steps, supported by their remarkably long toes and claws. While foraging, they turn over lily pads in search of snails, arthropods, and other small animals.

Plovers (Charadriidae) and the crab plover (Dromadidae) usually forage on open ground, relying on sight to locate the invertebrates on which they feed. The foraging bird runs a few steps, pauses with head cocked, then pecks at possible prey or runs again. Most plovers feed during the day, but the crab plover feeds mostly at twilight. Oystercatchers (Haematopodidae) feed largely on mussels, oysters, and marine worms. Depending on the type of mussel bed being exploited, the bird either tears loose a mussel and hammers a hole in the shell or, finding the mussel in water with its shell open, drives the knifelike beak into the open shell, cutting the adductor muscle and preventing the shellfish from closing.

Sandpipers and their relatives (Scolopacidae) use their slender bills as forceps to pick up surface invertebrates or (especially in the calidridine sandpipers) for probing in mud. Curlews use their long, downcurved bills for probing crustacean burrows on beaches and worm burrows on mudflats, and for picking up insects and berries on tundra or grassland. Woodcock (*Philohela* and *Scolopax*) also probe, but in woodland soils and leaf litter, feeding extensively on earthworms. Turnstones (*Arenaria*) habitually flip over vegetation, debris, soil, and stones with the straight upper edge of the bill, eating the animal life thus exposed.

Avocets and stilts (Recurvirostridae) feed in shallow water by sweeping the opened bill from side to side over the bottom (avocets) or near the surface (stilts). Stilts also feed by pecking and probing. The ibisbill (*Ibidorhyncha*), a recurvirostrid with a downcurved bill, inhabits Himalayan lakes and rivers and feeds by reaching under rocks in water for insects, mollusks, crustaceans, and worms, sometimes while wading belly-deep. Phalaropes (Phalaropodidae) habitually feed while swimming and sometimes stir up prey by spinning around.

Coursers and pratincoles (Glareolidae) are insectivorous birds of open country. Coursers feed most actively at night, taking termites, black ants, and other terrestrial arthropods by short dashes. Pratincoles hawk insects on the wing in the manner of swallows.

Sheathbills (Chionididae) inhabit Antarctic regions, where overdependence on one type of food may be disastrous. They subsist on algae; limpets and other mollusks; crustaceans; fish; the eggs and nestlings of penguins, cormorants and other birds; afterbirth and droppings of seals; and human refuse.

The food habits of seed snipe (Thinocoridae) are unique within the order. These chunky, terrestrial birds eat primarily vegetable matter, such as seeds, buds, shoots, and leaves, for which their short, stout bills are well suited.

Food habits of gulls and relatives

The gulls and their relatives (suborder Lari) are more dependent on flight for obtaining their food than are the shorebirds. The strongest fliers are the skuas and jaegers (Stercorariidae), which are gull-like in general proportions but with hooked bills. Stercorariids harass terns, gulls, boobies, and other seabirds until the latter drop or regurgitate food, which is retrieved by the "pirates." On their Arctic or Antarctic breeding grounds these birds prey on insects, rodents, small birds, and the eggs and young of other seabirds.

The diet of gulls (Laridae) is highly varied, including fish, small birds, and rodents, and a wide range of invertebrates, taken by active predation, as well as carrion of all sorts, garbage, and some vegetable material. Many of the larger gulls are not beyond piracy, stealing food from other birds (including members of their own species), and some prey heavily on the eggs and young of other seabirds. Sometimes they hawk insects or break open shellfish by dropping them from a height.

Most terns are smaller than gulls. All have straight, sharp bills and feed chiefly on the wing, by hovering over the water and plunging in for surface fish and crustaceans, by swooping low to pick fish from the water in flight, or by hawking insects over land.

Skimmers (Rynchopidae) feed by day or night. The bird flies over the surface of calm water at speeds up to 30 miles per hour with the long, knifelike lower mandible cutting the water. When the mandible strikes a fish or shrimp, the head doubles under the body and the bill clamps shut. The bird flies upward and swallows the fish in the air or carries it to the nest.

Food habits of alcids

Alcids (family Alcidae) are the only charadriiforms adapted for swimming underwater, which they do by propelling themselves with half-open wings. The larger alcids (puffins, murres, auks, and guillemots) feed on small fish and invertebrates, the smaller ones (murrelets and auklets) almost entirely on invertebrates, especially on tiny, free-swimming (planktonic) crustaceans, such as euphausids.

Breeding habits of shorebirds

Reproduction. The modes of reproduction in the charadriiforms are only slightly less diverse than the food habits. Nearly all are ground nesters, laying few eggs (two to four in most families; up to six in jacanas) often with little or no nesting material. Usually the eggs are protectively coloured. Some species nest in crevices or burrows, a few in trees. Many are colonial, with aggregations running to over a million pairs. Sex reversal, in which the female is more brightly coloured than the male and takes a dominant role in courtship, while the male incubates the eggs and rears the young, is found in most jacanas, painted snipe, and phalaropes.

Female jacanas are larger and more aggressive than males; after laying, females of some species show no interest in the nest or young, leaving incubation and care of chicks to the male. The nest is a shallow, sodden pile of vegetation that floats among aquatic plants. Chicks are downy and run well when a day old. Both chicks and adults may dive to escape danger and remain submerged (probably clinging to vegetation) with only the bill exposed. Adults have been known to take chicks up to 12 days old under their wings and carry them to a safer location. Jacanas perform injury-feigning displays and other striking displays, sometimes with their broad wings stretched aloft, and they attack other species near the nest.

The female of the Old World painted snipe (*Rostratula benghalensis*) is larger and more brightly coloured than the male, whereas the South American species (*Nycticryphes semicollaris*) shows little sexual dimorphism, although females are slightly larger. As in some jacanas, the female of *Rostratula* is more aggressive in courtship, displaying with spread wing and tail, and she leaves incubation and care of the chicks to the male. The nest is a shallow platform of bent reeds in which the four eggs (two in *Nycticryphes*) are incubated for about 19 days. The downy young take to the water readily.

Phalaropes nest in open tundra, marsh, or sedges near water. Females take the lead role in courtship and aggression, and several may vie for a male and fight among themselves. Copulation occurs on the water. There is no pronounced territorial behaviour around the nest, but a female may drive other females away from her mate. Incubation and care of the brood is performed by the male, and females leave the nesting grounds with the onset of incubation. The chicks are aquatic, swimming within an hour of hatching, and they perform the spinning habit commonly used by adults when feeding in water.

Oyster-catcher communal display

Oystercatchers (Haematopodidae) are noisy, stocky birds of coasts and shores. They are highly gregarious and even perform sexual displays within the flock. One such display is the piping performance involving a number of birds that run toward a central point, holding the bill and neck down and the bill open, uttering a loud piping. Groups also pipe and posture (lowering the bill and neck) in flight. Before copulation the male circles the female in a crouching pose with tail and bill lowered, while the female indicates readiness by raising her tail. All behaviour patterns (except the precopulatory ones) are common to both sexes. Oystercatchers have no strong territorial behaviour at the nest, often tolerating strange birds nearby, but they may return to the same spot with the same mate in successive years. Although pairs separate from the flock to nest, they periodically rejoin communal displays during breeding. The two or three eggs are incubated by both parents. Downy young move to the water's edge, where they remain on a feeding territory for at least six weeks and are fed small worms by the parents, who forage within sight in shallow water.

Avocets and stilts (Recurvirostridae) resemble oystercatchers in basic breeding behaviour, although the specific display patterns are different. Avocets nest in small colonies of up to a few hundred pairs, in grassy salt- or freshwater marshes that have substantial areas of shallow open water. Both sexes incubate the four eggs and protect the young. Stilts are somewhat colonial in their nesting, sometimes sharing a marsh with avocets. Stilts perform group displays in the breeding season. The ibisbill lays its eggs among the rocks of glacial streams. When alarmed, it bobs its tail and pumps its head up and down.

Most plovers nest in open areas, relying on their coloration to protect them while incubating or brooding. The solid colour of the back in the true plovers (*Charadrius*) matches the ground, and one or more black bands on the breast and face break the outline of the bird.

Many plovers are somewhat colonial but also exhibit territorial and aggressive behaviour. Some perform song flights during the breeding season, and many have melodious, whistled calls and bob when alarmed. The nest is a scrape or hollow with scant lining, in which four eggs are laid. Incubation is usually shared by the pair and takes from 20 to 30 days, depending upon the species. The chicks feed themselves but are usually protected by both parents by means of alarm calls that elicit crouching in the chicks. Often, parents will feign injury to divert a predator from their young. Many species occur in large postbreeding flocks and migrate great distances.

Crab plovers (*Dromas ardeola*) breed colonially in burrows in sand banks and fashion a nest chamber at the end of the narrow, three-to-five-foot tunnel. Here the bird lays a single, relatively large, white egg. The downy young is fed in the burrow by both parents.

Most pratincoles and coursers nest like plovers. Pratincoles breed in scattered groups, laying two or three eggs in an unlined scrape. They have a distraction display at the nest. The Egyptian plover (*Pluvianus aegyptius*) buries its egg in sand by day and incubates at night. Most glareolids lay two eggs, but the double-banded courser (*Rhinoptilus africanus*) lays only one, often located near antelope droppings, for concealment on otherwise bare ground. In that species, incubation by both sexes lasts about 26 days, and eggshells are removed. The chick has sparse down and is fed for about six weeks, until nearly fledged. A second egg may be laid before the first chick is independent. Adults and young combat the desert heat by panting, and adults also raise dorsal feathers and expose their legs during incubation.

The thick-knees frequent dry open ground but within reasonable distance of water. They are cryptically coloured and spend much of the day squatting on their heel joints ("knees") or resting flat on the ground with the legs drawn under the body. On hot days the legs may be extended behind the resting bird, perhaps for heat dissipation. They have large eyes and are active and noisy from dusk until dawn. Their displays include spread wing and tail postures. The one or two eggs are laid on bare ground and are incubated for 25–27 days by the female or by both birds. Both parents attend the downy chicks, which run from the nest a day after hatching and fly when six weeks old.

Sheathbills are the most gull-like of the Charadrii. Their feet are not webbed, but they swim well and fly strongly at sea hundreds of miles from land. Males bow, bob, strut, and coo during courtship. They build bulky nests in holes, burrows, or rock crevices. Their two or three eggs are laid more than a week apart, but incubation begins with the first. As a result, the chicks are of different sizes and rarely does more than the largest one survive. This seemingly wasteful system is an adaptation that allows the sheathbills to fledge the maximum number of young permitted by their fluctuating food supply.

Most members of the Scolopacidae (the largest family of the shorebirds) construct nests on the ground, but several species use the abandoned nests of other birds, in trees or on the ground. Most of the calidridine sandpipers have courtship songs given in flight and are monogamous, but the ruff (*Philomachus pugnax*) is notable for social courtship, performed on the ground, and for promiscuity. Males have a prominent collar of feathers of the head and neck (the "ruff") that are of different colours and patterns in different individuals. They assemble at communal display grounds (the arena or lek) where each bird occupies and defends a site during the day; occasionally, an "apprentice" male may be permitted to share an occupied site. Females visit the lek briefly and copulate with one or more birds. There is no pairing, and females assume all domestic responsibilities.

Woodcock and snipe (subfamily Scolopacinae) nest on the ground in dry woods (woodcock) and marshes (snipe). The males of most species perform courtship flights in which some or all of the "song" is produced by specialized wing or tail feathers. Woodcock are largely solitary and nocturnal and are known to fly to safety with a chick carried between their legs.

The subfamily Tringinae is the most diverse of the subgroups of the Scolopacidae. The breeding behaviour of the greenshank (*Tringa nebularia*) includes many features common to a number of shorebirds and provides a useful model for comparison with the behaviour found in the Lari and Alcae.

Breeding behaviour of the greenshank

From their winter quarters throughout much of the Old World temperate regions, greenshanks usually return north each year to the same breeding territory and often to the same mate. Males usually precede females to the breeding grounds, but sometimes they arrive together. Their nesting grounds are in northern Europe and Asia, in flat meadows or swamps near lakes, or in bogs with clusters of trees. The nesting territories vary in size from 100 to 700 acres, and nests may be three to eight kilometres (two to five miles) apart, or as close as a half kilometre. When the territories are relatively small, singing and fighting reaches a high pitch. In aggressive displays, opposing males lean forward with tails fanned, meeting bill to bill. They may then flutter over one another with legs dangling, in leapfrog fashion. Potentially dangerous intruders are met with violent scolding calls, flicks of the open wings, or shaking the half-raised wings. Anxiety is expressed by bobbing and curtseying or stretching the neck.

Flight songs are performed frequently during courtship and territorial adjustment, intermittently throughout incubation, and less regularly later. The tempo is quickest in unmated birds, which fly high (sometimes out of sight), singing, on downward glides and upward flight, while soaring and circling, sustained bursts of "too-too-too," with other calls interspersed. Unmated males chase other birds in the air. Upon landing the male gives an aggressive display toward other individuals. Males return the display, but females fly off, with the male in wild pursuit, sometimes attaining a great height and ending in a spectacular dive. The female may glide on bowed wings while the male rises and dips over her, dangling one or both legs, or he may flutter over a standing female in leapfrog display. The female invites copulation by tilting forward and squatting, and the male approaches and mounts, sometimes while waving his wings. Copulation is attempted repeatedly by the male before the female becomes receptive.

The male escorts the female to various parts of the territory, but she chooses the nest site and makes a scrape by lowering the breast, turning, and scratching with her feet; she also brings nearby plant material in her bill. A former nest may be reused for three or more years. The site is usually close to a landmark such as a boulder, stump, or tree. On occasion a single male greenshank may command two females that nest four to 250 metres apart or that occasionally lay in the same nest, where they may incubate side by side.

A clutch of four speckled and cryptically coloured eggs is laid in late April or early May and is incubated for 23 or 24 days. The eggs are laid at any time of day, and the female is accompanied by the male to the nest during the laying of at least the first eggs. Both members develop incubation patches (featherless areas of skin abundantly provided with blood vessels) on the sides of the breast and across the abdomen, and both share in incubation, although occasionally the male defaults. Duty shifts are most frequent at early morning and evening hours, and males generally incubate at night. The free individual spends its time at feeding grounds—usually a lake or marsh up to eight miles away, where it feeds, preens, or rests, or it may engage in song flights over the territory. The incubating bird sits tightly, even in the presence of danger, but it may spring off the eggs and perform an injury-feigning display. An egg knocked out of the nest is drawn back in by rolling it with the underside of the bill.

When the eggs are pipped the male spends much time perching near the nest. Chicks pipe inside the eggs and the hen calls back; she may or may not enlarge the egg hole by pecking. Immediately after the hatching the adult carries the shells some 50 to 500 metres and drops them, or occasionally eats them. The chicks, covered with protectively coloured down, leave the nest during their first day, running rapidly on relatively large feet. They scatter and squat at the alarm call and rally to gutteral clucking by the parent. The female herds the chicks toward the lake, preceded by the male, who acts as sentry. Chicks may have to cross streams or rivers and survive cold rains, but they are brooded during storms and at night for about two weeks. They are sometimes taken by weasels, gulls, hooded crows, or other predators. Sometimes the family separates, each parent taking one or more chicks, or one parent may depart, leaving all responsibility to the other. The chicks are feathered and fly at about 26 days of age but stay near the parents until they are experienced fliers.

Breeding behaviour of herring gull and relatives

The members of the suborder Lari are quite different in breeding behaviour from those of the Charadrii. The herring gull (*Larus argentatus*) is typical of many of the better known gulls. It inhabits subarctic and temperate regions of the Northern Hemisphere. It is not strongly migratory, but most birds shift southward after breeding, and some go as far as Panama, the Hawaiian Islands, the central African coast, northern India, Southeast Asia, and the Philippines.

In winter herring gulls have a daily routine of flying from their social sleeping grounds to the feeding grounds, near which they may rest and preen, returning to the roosting area in the late afternoon. Storms may flood the broad beaches or sandbars that they require for roosting, forcing them to shift many miles in search of a new area.

As spring approaches, the birds drift north to their breeding grounds, over which they may wheel one day only to disappear and return on another day. Eventually the birds land on the dunes or rocky coast that has been their traditional colony. Some stand in pairs, while others are in small groups known as "clubs." Old birds have already paired in February, and they return to their former territories, which are about 30 to 50 metres in diameter. The gulls establish or defend their territories by a number of aggressive displays, which may occasionally lead to fierce fighting.

Threat display of one bird may be met by threat, by anxiety or appeasement postures, or by retreat. The individual is usually dominant on its established territory. One aggressive display is the "oblique with long call," in which the neck is stretched obliquely forward and upward, the wings held slightly out from the body, and a loud, long-drawn call given with the bill wide open. In the silent "upright" threat posture the head is raised, the bill angled downward, the wings held stiffly apart from the body. A bird that approaches an intruder in this posture may increase its speed to a charge, half running and half flying. The intruder usually retreats to his own territory, where he, in turn, will perform the upright threat posture. Another aggressive action is "pecking the ground," in which the bird pecks and often tears out moss or grass, which it holds or tosses aside. Tugging at grass is strenuous and occasionally results in a backward tumble. A third threat posture, called "choking," often used by both members of a pair to intimidate another pair, consists of standing with legs bent, breast lowered, head pointed down, and the tongue lowered, while rhythmically jerking the head and uttering a deep call. These displays are effective in intimidating intruders, but occasionally fights break out in which birds peck or hold each other with their bills and pummel each other with their wings. Threat and fighting subside after territories have been established, and groups of neighbours recognize and tolerate each other.

Younger birds, not yet paired, settle in clubs. Here they doze or preen, and sometimes threaten or chase newcomers. Young males perform the "oblique with long call" display, causing other males to avoid them and females to land near them. Females approach males with the neck drawn in and the body and head horizontal, sometimes "head tossing" and uttering a liquid call. Instead of threatening, the male may be induced to walk off with such a female. Aggressive feelings are lessened by a display called "facing away," in which one or both birds, with neck stretched up, abruptly turn the head to the side away from the other bird. Together they make incomplete nest-building movements in a posture that resembles choking. The male may regurgitate half-digested food, which the female takes from his mouth and eats. This courtship feeding is repeated during pair formation but becomes less frequent as copulation becomes more frequent. Copulation is preceded by head tossing and begging calls by both members. The male stretches his neck and mounts the female's back while uttering a hoarse rhythmic call. The female continues to toss her head or may reach up to touch the male's breast while he waves his wings to maintain balance.

Courtship feeding

Copulations, choking, and courtship feeding may occur at the club, but eventually the pair walks off to establish a territory. They make several hollows (scrapes) with their feet, and begin to build a nest at one of them. Both collect bits of straw and moss, which they bring to the nest and which they may deposit while sitting in the nest with a sideways movement of the head. They scrape and turn in all directions to shape the nest.

The female lays three eggs at two-day intervals. With the first egg one bird (or both) stays at the nest, sitting or standing guard. They incubate in turn, relieving each other after periods of about two to five hours and often bringing new nest material. The arriving bird may call or "choke" and may have to push its reluctant mate off the eggs. If an egg is knocked out of the nest the bird rolls it back in, precariously balancing it with the thin lower edge of its bill. An incubating bird recognizes its returning mate by call even amid the clamour of the colony. Herring gulls apparently do not recognize their own eggs but return first to the nest site, even if the eggs are placed outside nearby. Incubation lasts about 30 days, including the three-day hatching period after the egg is first cracked. As part of the overall adaptations for concealment of eggs and chicks, gulls defecate at some distance from the nest and carry off the eggshells at hatching.

Defense of the brood lasts several weeks and consists of swoops at intruders, with one or both feet lowered. The alarm call of one bird in the colony alerts the others, but gulls learn not to react to alarm calls of individuals that tend to give false alarms. The chicks are brooded for about three days. They soon peck at the red spot on the adult's bill, which induces the parent to regurgitate food (usually worms) that is held for the chick to eat. Other activities of the chicks are preening, yawning, stretching, scratching, and crouching at the alarm call; later they run from the nest before crouching. Chicks that wander near other adult birds may be pecked to death.

After the chick's juvenile feathers have replaced the down and the bird approaches adult size and shape, its own parents may react aggressively toward it. The chick then adopts a submissive posture—with head withdrawn and body horizontal—similar to that of a female approaching a prospective mate. Full-grown young beg for food by head-tossing and high-pitched calls, but they are ignored or rejected as they become independent.

Herring gulls have concealingly coloured eggs, and their nests are widely spaced. The mobbing reaction in which many birds may participate is usually effective in distracting or repulsing a predator before it can extensively damage the colony.

The elaborate display repertoire of the herring gull during the breeding season is duplicated in other species, with some variations, additions, or deletions. The "hooded" gulls, exemplified by the black-headed gull (*Larus ridibundus*) and laughing gull (*L. atricilla*), have a striking "swoop-and-soar" aggressive flight display, and a ground display (called the "forward") wherein the neck is lowered, the head withdrawn and angled upward, and the wings held out from the body.

Several species of gull nest on narrow cliff ledges where they gain protection from predators but have limited space and face constant danger of the eggs or chicks falling off. In the kittiwake (*Rissa tridactyla*) adaptations against predation are reduced—alarm calls are rare, chicks are not camouflaged, defecation occurs on the nest rim, and eggshells are not carried away. Security against falling is achieved by a deep nest cup, smaller clutch (two eggs), and strong claws. The chicks crouch in the nest rather than running out.

The kittiwake is pelagic after breeding, and the swallow-tailed gull (*Creagrus furcatus*) of the Galápagos is semipelagic and nocturnal. The gray gull (*Larus modestus*) flies inland to waterless Chilean deserts to breed.

Breeding behaviour of terns

Terns are, in general, smaller than gulls. Most are coastal or pelagic during the nonbreeding season, returning to breed in large colonies on islands, offshore bars, isolated beaches, or on Arctic tundra. The common tern (*Sterna hirundo*) and royal tern (*Thalasseus maximus*) and their relatives nest on the ground, whereas noddies (*Anous*) nest in bushes or on cliff ledges, and the fairy tern (*Gygis alba*) deposits its single egg on the limb of a tree or bush. Nests of *Sterna* and *Thalasseus* may be so closely spaced that neighbouring birds spar with their bills as they incubate.

Displays of terns are more aerial than those of gulls and include "fish flights" in which one bird postures while carrying a fish. Ceremonial transfer of fish occurs during displays on the ground that are important in courtship and for maintenance of the pair bond.

Unlike gulls, most terns do not regurgitate food for the chicks but bring a fish back to the colony; there it is fed to the chick, which energetically pecks at the adult's bill. Chicks of royal terns band together soon after leaving the nest and may take to the water when predators approach. Terns are fierce in their mobbing attacks on predators. Like gulls, they often peck and kill chicks that trespass on their territories. Sooty terns (*Sterna fuscata*) have attracted considerable attention from biologists because on Ascension Island in the South Atlantic they breed every 9.6 months and on Christmas Island, in the Pacific Ocean, every six months. Elsewhere they have an annual cycle.

The breeding pattern of skimmers is ternlike. They breed on sandbars in rivers and estuaries in tropical and subtropical regions, forming loose colonies of about 100 to several thousand pairs. The nests are mere hollows in the sand. The eggs usually number three or four and are incubated chiefly by the female. The chicks are fed by both parents, mainly on small fish. In the presence of danger the chicks lie flat or burrow into the sand and may even kick sand onto their backs. Their mandibles are of nearly equal length until after fledging, enabling them to pick up fish brought in and dropped by the parents.

Although not strongly colonial in their nesting habits, numbers of skuas (*Catharacta skua*) may nest on the fringes of penguin colonies; in Iceland they nest in clumps of vegetation on great outwash gravel plains. Jaegers (three species of *Stercorarius*) nest on the tundra, where both sexes share in incubating the two eggs. Some species have light and dark phases. They are pelagic and solitary after breeding.

Breeding behaviour of alcids

Alcids breed in island colonies along Arctic and north temperate seacoasts, with the exception of a few murrelets that breed inland on mountains. Even these must remain within flying distance of the sea. The breeding behaviour of the pigeon guillemot (*Cepphus columba*) is fairly typical of the family. This species breeds on islands and coasts of the North Pacific, south to central California. It nests between rocks or in holes in cliffs, uses burrows of other birds, or digs its own tunnels with its bill and feet. Occasionally it nests on an open ledge or in tall grass. Unlike some other alcids, guillemots walk quite well on land.

Birds land at their breeding colonies in British Columbia in early April, at which time they stay only briefly in the early mornings. In general, birds return to the same mate and nest burrow for many years. Full attendance is achieved by the end of June, but even then many birds leave the colony in the afternoon to go to their feeding grounds—shoals that may be several miles distant. Only incubating and brooding birds spend the night at the colony. They drive other birds away from their burrows before egg-laying but are more tolerant later in the season when feeding young. The pair also defends a perch site away from the nest burrow, usually a rock close to the water. On this site, which is used in successive years by the same pair, copulation occurs and the nonincubating bird may rest and preen. About 30 percent of the birds at the colony consists of nonbreeding individuals—yearlings or two-year-olds that were hatched at the colony but are unpaired.

Communal water displays

During the pre-egg stage, which lasts about 30 to 60 days, the mated pairs spend much time at the perch sites, but they occasionally visit the nest site or cliff top and join in communal water displays. Their display repertoire includes several alarm reactions—the flight intention call, bill dipping (in water), and a scream delivered with the neck straight up and the bill agape. Among the aggressive and sexual displays are a silent lunge on water or land with bill somewhat open, chases that may be in flight or underwater, and a "hunch-whistle" on shore or water, with tail cocked, the head drawn back, the bill agape, accompanied by piping. A display that may represent appeasement is the "twitter-waggle" in which the bird twitters with the tail raised, wings loose, and head and neck outstretched and wagging sideways. The communal water dances, which occur in the pre-egg stage, involve many birds. Pairs are not always together or both present. The birds move rapidly under water or at the surface propelled both by wings and feet. This behaviour is interspersed with "hunch-whistles" and "twitter-waggles."

Pair formation is indicated by mutual "billing," in which the birds waggle their heads but rarely touch bills, while uttering a twittering trilled song with the bill open, revealing the bright red mouth lining. Copulation is initiated by billing, followed by the male's waddling in a circle, first in one direction, then in the other. The female crouches and the male mounts, resting on his tarsi and fluttering his wings to keep balance; the female may gape, scream, or utter an alarm call. She then rises, throwing off the male, and the birds bill or preen.

The pair cleans debris out of the nest cavity and piles up pebbles, without making a well-defined nest. A normal clutch is two eggs (rarely one), and the birds' brood patches will not accommodate more than two eggs. The second egg follows the first by three days. Incubation lasts for about 32 days, including a two-day interval between cracking and hatching. Steady incubation during the day begins one day after the clutch is complete, but incubation at night may not begin for one or two days more, at which time a single bird begins to sit all night until the early morning exchange. The incubating bird may leave the burrow briefly to defecate, preen, bathe, or drink, but it will not go to the feeding grounds until its mate arrives at the burrow. The downy chicks are fed, beginning the day after hatching, by both parents during their 35-day nestling and fledging period. Their food is mostly fish, including many blennies, sculpins, and sand lances, brought in the parent's bill throughout the daylight hours. The chicks may reject fish that are awkward to swallow or that have sharp spines. Fledged chicks eventually desert the burrow by night and may be coaxed out by a fish dangling in the parent's bill. The young are then independent and they disappear offshore to feed without the parents, who may remain to loaf about the colony during morning hours for several weeks.

Behaviour patterns like those of the pigeon guillemot run through many species with various modifications. Many alcids are burrow nesters, and some (*Uria*) habitually nest on narrow ledges of vertical cliffs, where they rest on their tarsi facing the rock wall. Some murrelets (*Brachyrhamphus*) nest in rock crevices above the timberline on mountains, and in trees of the taiga zone. Both sexes incubate the one or two eggs and feed the young at the nest site until it is nearly or completely fledged. Other murrelets (*Synthliborhamphus*) are truly nidifugous (precocious), their downy chick taking to the sea within two days of hatching. Many alcids nest in immense colonies of mixed species, which segregate in part according to their specialized needs for nesting and feeding. They migrate to pelagic winter feeding grounds by swimming.

FORM AND FUNCTION

Most members of the Charadriiformes are clearly recognizable as belonging to a particular suborder. Major structural variations occur in the beak and legs, correlated with the mode of feeding and size of food.

Adaptations for flight. Gulls, terns, and skimmers have long, narrow wings, low wing loadings (the ratio of weight to wing area), slow wing beat rates and flight speeds, and moderately developed flight muscles. Alcids, on the other hand, have proportionately shorter wings, high wing loadings, rapid wing beat rates and greater speed (45–55 miles per hour), and large flight muscles. Their wing bones are flattened in adaptation to underwater "flight." The shorebirds lie between the other suborders in flight adaptations; their wings are not adapted for soaring, but they are strong fliers and cover great distances in continuous flight.

The evolution of the wing in the Alcidae has been influ-

Drawing by G. Tudor

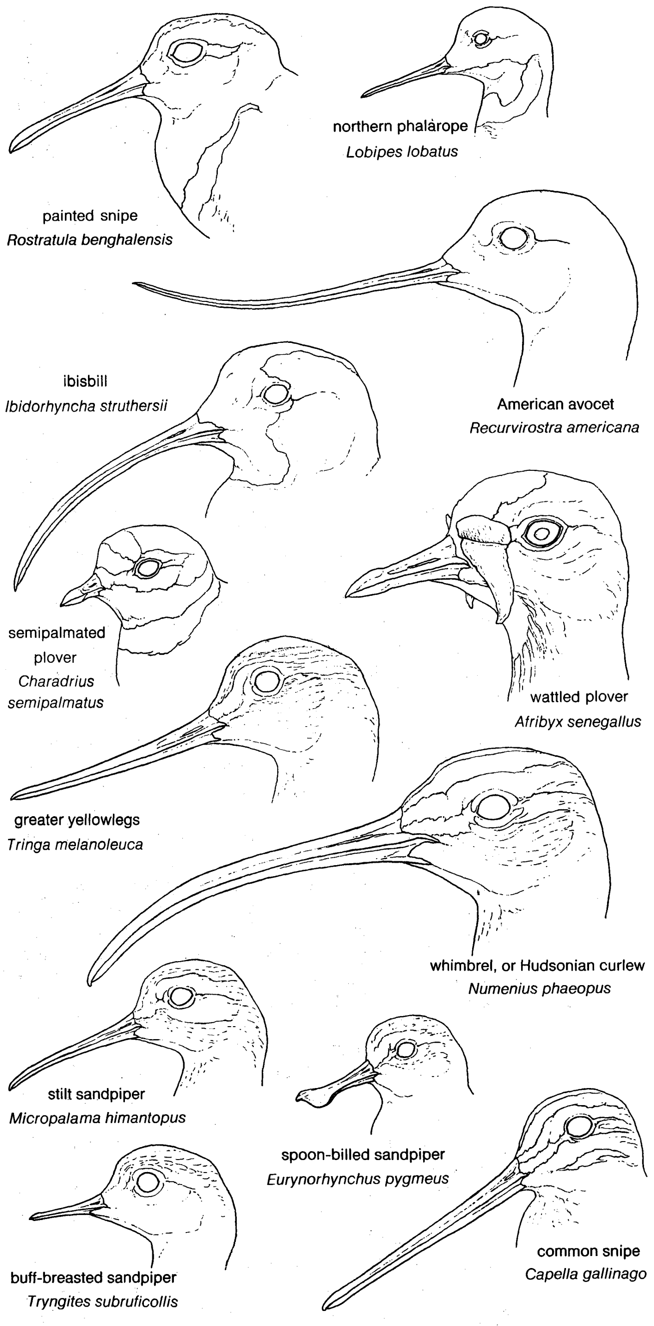


Figure 29: Bill variations among several shorebirds.

Aquatic use of wings

enced by the fact that it serves both as an aquatic paddle and as an aerial wing. The wing is partly folded underwater, reducing its area and increasing its mechanical advantage. It is used not to provide lift but to propel the bird in pursuit of prey. To support a bird in air, however, a wing must be larger than the optimum paddle size. This is especially critical in large birds, as body weight is proportional to the cube of linear dimensions and wing area to the square. Body size and flying ability are therefore limited by the need for a small effective paddle. In flightless aquatic species the wing can be relatively small and the bird can be much larger, as in the great auk and in penguins.

Molt. Most Charadriiformes have two molts between breeding periods, which are on an annual cycle in all except some populations of the sooty tern (*Sterna fuscata*). A partial body molt generally precedes breeding, and a complete molt follows breeding. In some species flight feathers of the wing and tail are molted before the fall migration; in others they are retained through the fall migration and molted on the wintering grounds. Still other species stop along the migration route to molt. The ruff (*Philomachus pugnax*) is exceptional in having one complete and two partial molts between breeding cycles.

Major flight feathers of most species are lost gradually and in sequence so that flight is not impaired. Flight feathers of the wing in most Alcidae, however, are molted simultaneously after breeding, with temporary loss of flight but no hindrance to underwater locomotion. One jacana (*Actophilornis*) is also flightless for a brief period.

Adaptations for feeding. As in other birds, the upper jaw as well as the lower jaw can be moved up and down. The lower jaw has special regions of flexibility enabling it to be widened by bowing outward and allowing passage of large prey into the throat in Lari and Alcae.

The shape of the bill varies greatly within the order, in accordance with special feeding methods. The hooked tip in gulls and jaegers facilitates grasping and tearing of food, which cannot be managed by the straight-billed terns. Some species such as the Atlantic puffin, rhinoceros auklet, and fairy tern are able to carry several fish crosswise in the bill while capturing still more. They do this by holding the fish against the roof of the upper jaw with the spiny tongue. The shorebirds exhibit a wide variety of bill types (Figures 29 and 30). Most bills in this group are long and slender, and straight or curved up or down. In the Scolopacidae, the bill is usually slender and flexible; in the Charadriidae, somewhat stouter and less flexible, often slightly swollen at the tip. Among the specialized probers (sandpipers, snipe, and allies), the upper jaw is rigidly attached to the cranium and has a mobile tip with a concentration of tactile sense organs under the rhamphotheca (the horny covering of the beak). The tip of the upper jaw is controlled by jaw muscles and serves to grasp worms or larvae underground and to inch them along the bill before swallowing.

In one plover (the wrybill, *Anarhynchus frontalis*) the bill curves to the right; in the spoonbilled sandpiper (*Eurynorhynchus pygmeus*) the tip of the bill is broad and spatulate. Oystercatchers have laterally flattened bills, with the tips forming a vertical blade. Seed snipe have short, conical bills not unlike those of sparrows. Pratincoles have short bills and wide mouths, like those of swallows and swifts.

The Lari show few bizarre modifications of the bill, feathers, or feet. One oddity is the bill of skimmers, in which the upper mandible is streamlined in cross section and laterally compressed. The lower mandible is knifelike and protrudes well beyond the upper. The sharp upper edge of the lower mandible fits into a groove in the upper when the bill is closed (see Figure 31).

Many alcids acquire bill ornamentation or head plumes during the breeding season. The functions of elaborate bill modifications are poorly understood but are believed to involve courtship and species recognition. The highly coloured plates on the bills of puffins (*Lunda* and *Fratercula*) are shed when the birds molt following the breeding season. In alcids that feed on zooplankton (tiny, free-floating animals), such as the least auklet (*Aethia pusilla*) and the dovekie or little auk (*Plautus alle*), the bill is relatively wide, the tongue large and fleshy, the palate broad with numerous horny projections, and the throat provided with an expandable pouch.

Many charadriiform birds drink saltwater. The ionic balance of their blood is maintained not only by the kidneys but by supraorbital glands (called nasal or salt glands) that lie in grooves in the skull over the eyes, discharging their salty excretion through the nostrils.

EVOLUTION AND PALEONTOLOGY

The earliest known Charadrii are represented by miscellaneous bones from the Upper Cretaceous rock layers of Wyoming, considered to constitute a family, the Cimolopterygidae, of the Charadrii. These bones, comprising two genera and four species, show closest resemblance to the avocets. Of similar age, but extending into middle Eocene times, are the Paleotringinae of the Scolopacidae, known from fossil deposits in New Jersey and Wyoming. These consisted of five species in a single genus of poorly known, large shorebirds, represented by fragmentary limb bones. Other probable relatives of the avocets, the Presbyornithidae, appear in the lower Eocene rocks of Utah. They are thought to be heavier bodied and more aquatic than present-day avocets, and to represent an ancestral stock from which the avocets and stilts descended. An extinct genus of the Rostratulidae and the modern genus *Limosa* are reported from the middle and upper Eocene of Europe. In the Oligocene are found extinct species in the modern sandpiper genus *Tringa,* in the plover genera

Drawing by G. Tudor

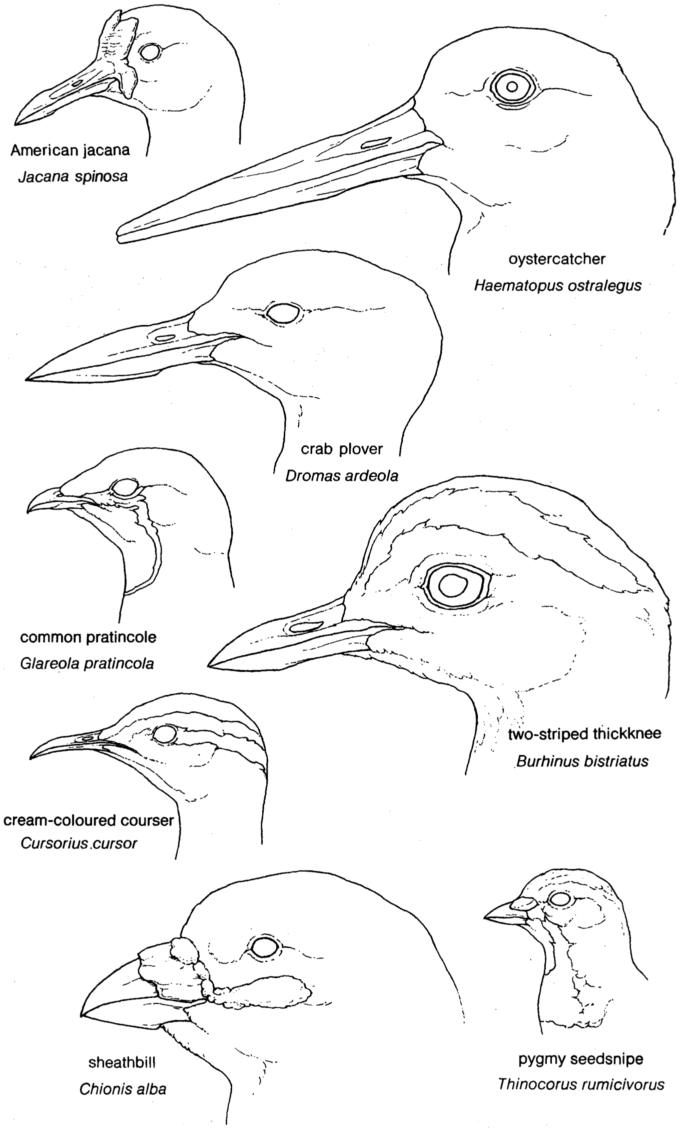


Figure 30: Variations in the bills of shorebirds.

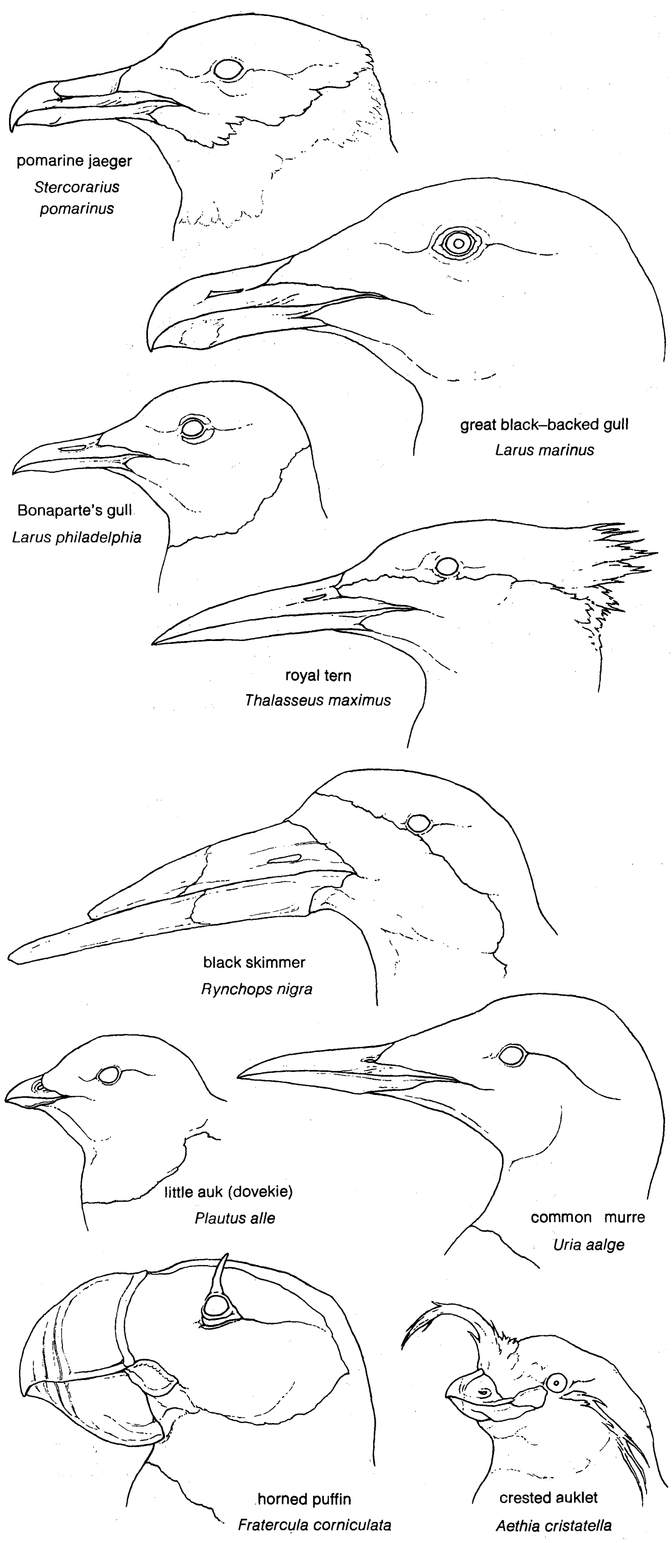


Figure 31: Variation in the bill among Lari and Alcae.
Drawing by G. Tudor

Charadrius and *Vanellus,* and in extinct genera of the Scolopacidae and Charadriidae. Possible relatives of the jacanas were the Rhegminornithidae from lower Miocene times of Florida; features of the tarsometatarsus suggest that these birds had a large hind toe. Other discoveries from the Miocene are extinct genera of the Scolopacidae, Charadriidae, Haematopodidae, and Burhinidae, and extinct species of four contemporary genera. Pliocene finds include the Recent *Bartramia* and *Micropalama* and an extinct genus of oystercatcher. Pleistocene deposits have added extinct species of four living genera, extinct genera of Scolopacidae and Charadriidae, and many examples of species still living today.

Within the Lari, extinct genera of the Laridae are reported from the Paleocene, Oligocene, Miocene, and Pleistocene, and an extinct genus of tern occurred in the Pleistocene. *Larus* and *Sterna* first appear in Miocene deposits—*Stercorarius* in the Pleistocene. Numerous modern species of Lari have been found in Pleistocene deposits.

The earliest of the auklike birds known at present were members of the alcid subfamily Nautilornithinae from the lower and upper Eocene of Utah. Represented by wing and leg bones, these birds are inferred to have had longer limbs than present forms and to have been less well adapted for flight underwater than are contemporary species. The Miocene produced the first known species of *Uria* and *Cerorhinca,* and an extinct genus of alcid. The murrelet genus *Brachyrhamphus* and the auklet genus *Ptychorhamphus* first appeared in the Pliocene, along with an extinct genus. Fascinating finds from the Pliocene are some very flattened wing bones (humeri) that exhibit features at both ends resembling those of penguins. These birds, the Mancallinae, were undoubtedly flightless and more specialized for underwater wing propulsion than was the now extinct great auk.

From this record of fossil Charadriiformes it appears that considerable adaptive radiation (diversification) has occurred and disappeared in the past. Modern families probably had their origins in Paleocene times or earlier, and many modern genera were well represented by the Oligocene and Miocene. Modern species date chiefly from the Pleistocene, and many facts concerning the distribution and differentiation of northern subspecies and species are explainable in terms of isolation during the periods of Pleistocene glaciation.

CLASSIFICATION

Features used in classification. The order Charadriiformes can be characterized by a variety of anatomical, behavioral, or other features, not all of which necessarily apply to any one species, genus, or family, and no one of which is necessarily found throughout the order. One set of characters may evolve rapidly in one group, another set in another. It is impossible to reflect all of these relationships in a single classification, and one must expect continued change in any classification with increasing knowledge of living birds and of fossils.

Annotated classification. The classification below follows that of the American ornithologist Alexander Wetmore. Groups indicated by a dagger (†) are known only as fossils.

ORDER CHARADRIIFORMES

Palate schizognathous; upper jaw schizorhinal (except Burhinidae and *Pluvianus*); lachrymal bone fused to ectethmoid; cervical vertebrae 15 or 16; dorsal vertebrae opisthocoelous or heterocoelous; two carotid arteries (except *Synthliborhamphus*); oil gland feathered; syrinx (vocal organ) tracheobronchial; wing diastataxic (except *Philohela*); aftershaft present; down feathers present on adults. Eggs usually pyriform (*i.e.,* somewhat pointed at one end). Chicks downy and nidifugous, pseudonidifugous, or nidicolous. About 312 species; worldwide.

Suborder Charadrii

Hypotarsus complex (with canals), coracoid bones usually separate, depressions for supraorbital grooves usually small or absent; basipterygoid processes and occipital foramina usually present; furcula without hypocleideum; adult downs on pterylae only. Young nidifugous (precocious).

†*Family Cimolopterygidae*

Fossil only. Known only from Upper Cretaceous strata of Wyoming. Four species; medium-sized shorebirds.

Family Jacanidae (jacanas)

Small to medium birds with showy plumage, moderately long, straight bills, and long legs. Some with wattles or lappets, a horny forehead plate, and wing spurs. Extremely elongated toes and long, straight claws. Occipital foramina and supraorbital grooves lacking; coracoids overlapping; rectrices 10. Females larger than males. 7 species; worldwide tropical and subtropical; length about 16–53.5 cm (6–21 in.).

†*Family Rhegminornithidae*

Fossil only; lower Miocene strata of Florida. One species; medium-sized birds.

Family Rostratulidae (painted snipe)

Small birds with cryptically patterned plumage, long slender bill, and moderately long legs. Female larger and brighter than male. Occipital foramina present; sternum narrow, with single pair of notches. Crop present in *Rostratula;* trachea convoluted in female. Two genera; one in southern South America, one in Africa, southern and eastern Asia, Australia; length 19–24 cm (7½–9½ in.).

Family Haematopodidae (oystercatchers)

Medium-sized birds with black, brown, or white plumage in bold patterns, or solid blackish brown. Bill long and wedge-shaped, bright red; legs moderately long, stout. Three toes. Supraorbital grooves large; tarsus covered with small, hexagonal scales. Four species, inhabit most temperate and tropical seacoasts, and inland water bodies in Europe and Asia; length 38–51 cm (15–20 in.).

Family Recurvirostridae (avocets, stilts, and ibisbill)

Moderately large birds with long bills, legs, and necks. Plumage in bold, simple patterns of black and white, gray, chestnut, or buff. Bill straight, recurved, or decurved. Toes webbed in *Recurvirostra.* Legs covered with reticulate scales. Plumage of underparts dense. 7 species worldwide in temperate and tropical regions, one species in Himalayas; length 29–48 cm (11–19 in.).

†*Family Presbyornithidae*

Fossil only. Two genera of large shorebirds from the lower Eocene of Utah.

Family Charadriidae (plovers, lapwings)

Small to medium-sized birds. Mostly with bold (but often concealing) plumage patterns of solid blacks, gray, browns, and white; many with one or two chest bands. Some with wattles and wing spurs. Bill usually short, with a swollen tip. Legs moderately long to long, with reticulate scale pattern. Hind toe usually absent. About 61 species; worldwide; length about 15–40 cm (6–16 in.).

Family Scolopacidae (snipe, woodcock, sandpipers, turnstones, and allies)

Small to medium-sized birds, mostly finely patterned in buff, browns, chestnut, black, gray, and white. Bill moderate to very long and slender; straight, decurved, or recurved; one with spatulate tip. Legs short to long, usually with transverse scales front and back. Hind toe usually present and elevated. About 82 species; worldwide; length 12.5–61 cm (5–24 in.).

Family Phalaropodidae (phalaropes)

Small, densely feathered birds with straight slender bills, moderately long legs, lobed toes, and flattened tarsi. Females larger and brighter than males. Three species; Arctic to temperate regions of Northern Hemisphere; winter at sea and in South America; length 19–25 cm (7½–10 in.).

Family Dromadidae (crab plover)

Medium-sized bird of white and black plumage. Legs long and covered with reticulate scales. Bill strong, laterally compressed, and pointed. Nostrils pervious; basipterygoid processes absent; 15 cervical vertebrae; dorsal vertebrae heterocoelous. One species; coasts of Indian Ocean and southern Red Sea; length 38 cm (15 in.).

Family Burhinidae (thick-knees)

Medium-sized birds with cryptically patterned plumage of brown, gray-brown, black, and white. Bill stout, short to moderately long. Tarsus reticulate. Upper jaw holorhinal; basipterygoid processes and occipital foramina absent; coracoids overlapping. Large eyes. Nine species; temperate and tropical regions of Eurasia, Africa, and Australia; New World tropics; length 35–52 cm (14–20 in.).

Family Glareolidae (pratincoles and coursers)

Pratincoles short-billed, long-winged, with medium-long legs and forked tails; coursers have longer bills, shorter wings and tail, long legs. Plumage patterned in olive, brown, gray, chestnut, black, white. Legs have rectangular scales front and back. Occipital fontanelles absent; basipterygoid processes absent in adult; dorsal vertebrae heterocoelous; 15 cervical vertebrae. 17 species; Africa, southern Eurasia, and Australia; length 15–25 cm (6–10 in.).

Family Thinocoridae (seedsnipes)

Small to medium-sized birds, cryptically patterned in brown, buff, gray, white, and black. Bill short and conical, legs short, wings long and pointed. Nostrils operculate; crop present; vomer broad; basisphenoidal rostrum thick; no basipterygoid processes or occipital fontanelles. Four species; southern South America and north in Andes; length 17–28 cm (7–11 in.).

Family Chionididae (sheathbills)

White birds of moderate size, short legs, and stout bill, with horny sheath over nostrils. No occipital fontanelles or basipterygoid processes; large supraorbital grooves; thick plumage; short carpal spurs on wing. Two species; inhabit islands of extreme southern Atlantic and Indian Oceans; length 35–43 cm (14–17 in.).

Suborder Lari

Hypotarsus simple (grooved but without canals); coracoids in contact (except in Stercorariidae); supraorbital grooves large; basipterygoid processes absent (present but small in young); occipital foramina absent in adults; furcula with hypocleideum; adult downs on both pterylae and apteria; anterior toes usually fully webbed, hind toe absent or minute; young tardily nidifugous (leave nest when half grown).

Family Stercorariidae (skuas, jaegers)

Medium-sized to large birds of solid brown, or brown, white, and black plumage. Bill moderately long, stout, hooked, with a horny dorsal plate. Wings long and pointed; tail with elongate central feathers in some. Legs rather short, toes webbed. Ceca well developed; coracoids not overlapping. 4 species; Arctic and Antarctic; length 43–61 cm (17–24 in.).

Family Laridae (gulls, terns)

Small to large birds, with solid plumage patterns of white, gray, black or sooty brown. Legs, eyes, and bill often brightly coloured in red or yellow. Wings long and pointed; tails of terns often long and deeply forked. Legs moderate to short. Bill short to moderately long; stout and somewhat hooked in gulls, more slender and pointed in terns. Front toes webbed. Sternum with two pairs of notches, ceca small. Rhamphotheca simple (without horny plates). About 85 species; worldwide; length 20–76 cm (8–30 in.).

Family Rynchopidae (skimmers)

Medium-sized birds with long, pointed wings, short legs, and a large bill. Bill deep and extremely flattened, lower mandible knifelike and longer than upper. Plumage black or blackish brown above, white below. Bill red or yellow, sometimes with black tip. Front toes webbed. Pupil closes to vertical slit. Three species; irregularly distributed in tropical and temperate rivers, lakes, and seashores; length 37–51 cm (14½–20 in.).

Suborder Alcae

Large supraorbital grooves with intervening space narrowed to ridge; basipterygoid processes absent in adults; occipital fontanelles present; haemapophysis of dorsal vertebrae large; sternum long and narrow with long, rounded metasternum. Anterior toes fully webbed, hind toe absent. Wing bones flattened. Young downy, nidicolous or nidifugous.

Family Alcidae (auks, murres, puffins and relatives)

Small to large, dense-plumaged, short-winged, aquatic birds. Plumage black, gray, or brown, usually white below. Many have ornamental bill and head plumes. Legs set far back; tail short. 21 species; coasts of northern oceans; length 16–76 cm (6–30 in.).

Critical appraisal. The Charadriiformes are thought to be related to the Gruiformes and to the Columbiformes. Certain intermediate or aberrant families have been removed from or added to the Charadriiformes by various taxonomists. The Belgian ornithologist R. Verheyen considered the Thinocoridae with the Pteroclidae (sand grouse) as a suborder of the Turniciformes, which he placed between the Columbiformes (pigeons and doves) and Galliformes (pheasants and relatives). It has been argued on behavioral and morphological evidence that the Pteroclidae are charadriiform rather than columbiform, but most authorities retain them in the Columbiformes on anatomical grounds. Verheyen allied the Jacanidae with two gruiform families (Rhynochetidae, Eurypygidae) as an order related to the rails and Charadriiformes. P.R. Lowe united the charadriiform and gruiform birds into a single order—the Telmatomorphae—placing the Thinocoridae as a link between the gruiform and charadriine members. Other families that have been variously placed in the Charadriiformes or Gruiformes are the Otididae (bustards, see *Gruiformes,* above), Burhinidae, and Glareolidae.

With undue emphasis on similarities resulting from convergent evolution taxonomists at times have allied the Lari with the Procellariiformes, and the Alcae with other diving birds (penguins, diving petrels, grebes, and loons).

Within the widely accepted limits of the Charadriiformes as presented here there are different opinions about the rank (superfamily, family, subfamily, or tribe) of various groups and about their internal relationships. Pending further evidence from comparative anatomy, physiology, biochemistry, and behaviour, the major groups within the suborders are here given family rank. (R.L.Z.)

Columbiformes (sand grouse, pigeons, doves, dodoes)

The order Columbiformes comprises the pigeons, doves, dodoes, and sand grouse. The suborder Columbae embraces the extinct dodo and solitaires in the family Raphidae and the extinct and living pigeons and doves in the family Columbidae. The names pigeon and dove are synonymous and imply no biological distinction. The sand grouse are given a distinct suborder, Pterocletes. The pigeon family is a natural and homogeneous assemblage of about 285 species of readily defined birds, unique in producing, for feeding their young, a nutritive secretion from the crop wall. Pigeon's milk is similar in composition to mammalian milk and is also induced by the secretion of the hormone prolactin from the pituitary gland. Pigeons also are distinctive in their unusual manner of drinking, in which water is sucked in as a continuous draft, the process being assisted by muscular contractions of the esophagus, whereas other birds take a sip of water and then tip back the head to swallow. It is frequently stated that the sand grouse share this typically pigeon habit; in fact they suck and then tip back the head to swallow, repeating the process four to 10 times during any drinking session.

GENERAL FEATURES

Adaptive radiation

With the exception of some highly specialized ground-living forms, all pigeons are readily recognizable. They range in size from birds the size of a starling (the diamond dove of Australia) to some as large as a female turkey (the crowned pigeons of New Guinea). The skeleton and body form is usually unspecialized, enabling pigeons to feed and roost arboreally yet also collect food from the ground. Adaptive radiation has been either toward a more specialized arboreal life or toward ground-feeding forms, some of which (quail doves) convergently resemble partridges. Sand grouse also resemble gallinaceous birds, although they cannot run, and have become adapted to desert and semidesert conditions in the Afro-Asian region. From the rock dove man has bred the various races of domesticated dovecote pigeons, racing pigeons, and other fancy breeds, while the domesticated form of the African collared dove (the so-called Barbary dove, sometimes erroneously given specific status) has long been a popular cage and dovecote bird. Because many pigeons are extremely well adapted to grain-eating and grazing habits, some conflict with man's agricultural activities and are considered pests.

Distribution and abundance

Pigeons are virtually cosmopolitan, being absent only from the Arctic, Antarctic, and some oceanic islands. Five species have become extinct since the late 17th century, at which time the dodoes and solitaires also vanished. The best known example is that of the passenger pigeon (*Ectopistes migratorius*) of North America, which was remarkable for its extreme gregariousness, a factor that helped the early settlers to exploit it ruthlessly; it was exterminated by the end of the 19th century. With the exception of several Pacific island races considered rare or endangered, other species have spread and increased, particularly as a result of man's agricultural activities. Since 1930 the collared dove has spread 1,000 miles northwest from the Balkans.

IMPORTANCE TO MAN

Throughout the world, agricultural development has effectively provided a "super-habitat" for many seed-eating pigeons, enabling them to thrive and spread: the collared dove (*Streptopelia decaocto*) in India in relation to cereal production, the spotted dove (*S. chinensis*) in Southeast Asia in relation to rice paddy, and the laughing dove (*S. senegalensis*) throughout Africa, Arabia, and India, associated with native crops. Any species that can profit from agricultural expansion must be extremely well adapted—in a sense, preadapted—to such conditions and is likely to achieve pest status. Damage by the wood pigeon in Britain has been estimated at £2,000,000 to £3,000,000 annually, but it is unlikely that it justifies expensive remedial action

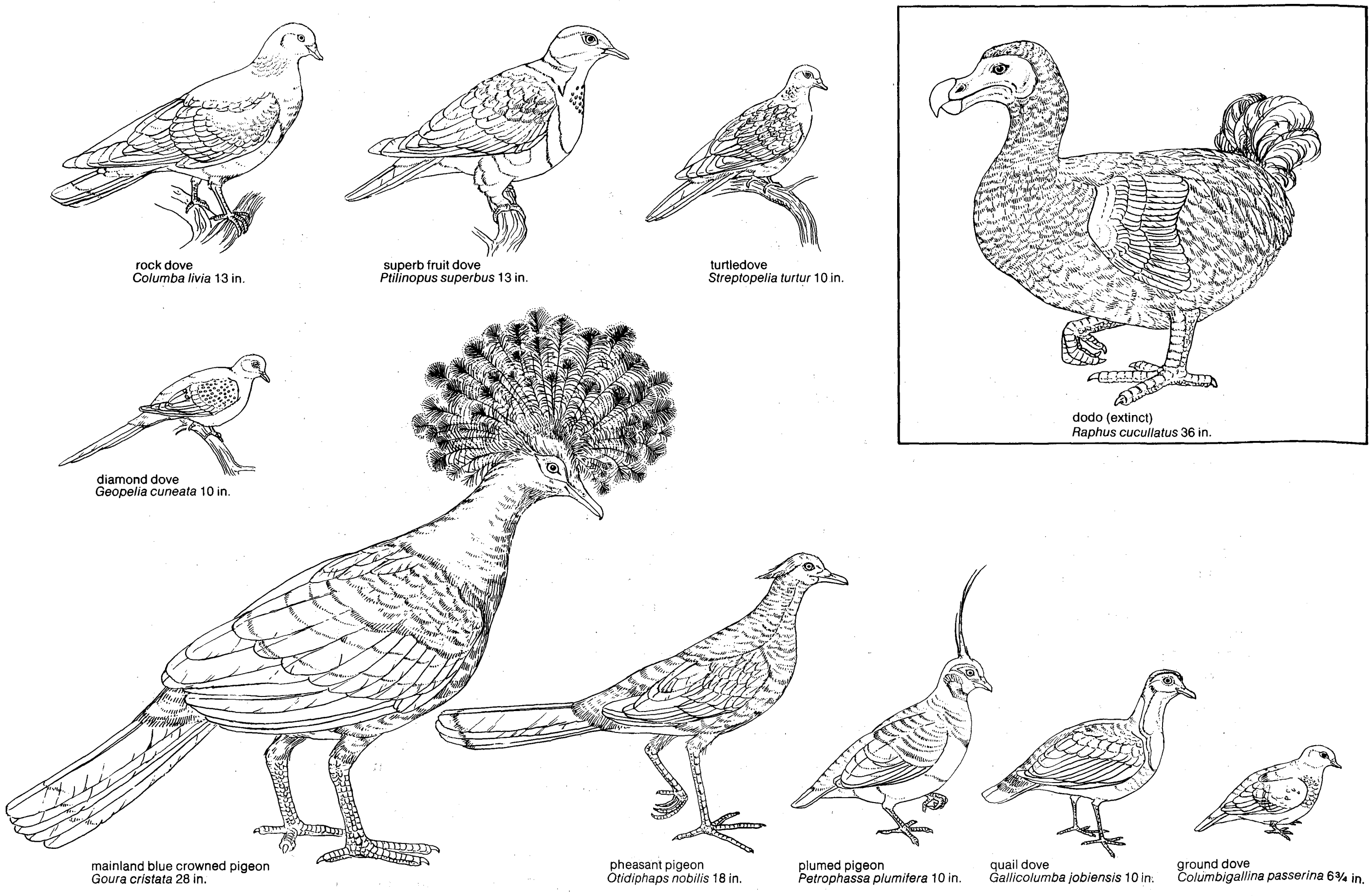


Figure 32: Types of columbiform birds.

on a national scale; the cost of cereal spillage before and at harvest exceeds this amount. Indeed, it has been shown that artificial population control, employing conventional methods such as shooting, is not feasible at a national level and that bonus schemes do not improve control operations. Emphasis is now placed on providing remedial action, relying often as much on scaring as killing, only in the precise locality where damage is occurring or is imminent. The use of more extreme control methods (for example, poisoning) is ruled out in densely populated areas because of the risks to people and livestock. In addition, the risk to other wildlife is now recognized, and there is a widespread desire, public as well as official, not to upset the balance of nature for the sake of immediate but marginal benefits. Any damage caused by the mourning dove to cereal farming in the United States pales in comparison with the value of the species as a sporting bird.

Nesting colonies of the rock dove (*Columba livia*) were farmed by Neolithic husbandmen for food, and gradually the process of rearing young in confinement led to the production of domesticated strains. Evidence for domestication extends back to 4500 BC in ancient Iraq, and the bird was sacred to the early Middle Eastern cultures, being associated with Astarte, the goddess of love and fertility; later, in ancient Greece, it was sacred to Aphrodite and in Roman times to Venus. By the Middle Ages dovecote populations were kept as a source of food on virtually every manorial estate in Europe. From the domesticated pigeons have been derived the various fancy breeds, such as tumblers and pouters, and many genetic aberrations that have given pleasure to countless enthusiasts. From the same source have come racing pigeons. Belgium, at the top of the international league table, has more than 165,000 pigeon fanciers. Carrier pigeons were used to relay news of the conquest of Gaul to Rome, brought news of Napoleon's defeat at Waterloo to England, and were used extensively for message carrying in the two world wars.

Urban pigeons

The popularity of the dovecote pigeon declined in the late 19th century as farmers realized that it was more efficient and brought greater financial reward to supply nations with bread than to convert corn inefficiently into this intermediate form of protein. The release of thousands of pigeons, together with escapes, established the feral populations in numerous European towns, in North America (where it is often known simply as the "city pigeon"), and other parts of the world as far away as Australia. Being naturally adapted to rocky ravines, sea cliffs, and barren sites, the bird has readily accepted the sides of buildings, bridges, and other man-made "cliffs." In towns it is fed and protected by an indulgent and benevolent public, at the same time being cursed by public health inspectors and those concerned with its depredations on stored grains. To a large extent problems connected with food storage can be remedied by adequate proofing of buildings and control of spillage.

The importance of feral pigeons as a reservoir and means of transmission of disease is becoming increasingly recognized, even though there are few cases where the transmission of disease can be proved. Feral pigeons appear to harbour ornithosis (psittacosis) to a sufficient extent to provide a potential human risk. Up to three-quarters of various local pigeon populations examined in Paris were found to be infected, and the virus also has been isolated in pigeons in Liverpool, London, and elsewhere. Virulent strains of *Cryptococcus neoformans,* which causes a fungal disease of the skin, as well as cryptococcal meningitis (a disease of the lungs and central nervous system), have been isolated in pigeon excreta in various cities.

NATURAL HISTORY

General habits. The evolution of the crop has been of vital importance to the pigeons. This bilobed diverticulum (a blind pouch) of the esophagus, located just posterior to the buccal cavity, serves as a storage organ for food. Subsisting for the most part on seeds, buds, leaves, and fruits, foods of low protein content and nutritive value, pigeons must consume large quantities during each feeding day. The rate of food intake can be higher than the digestion rate would otherwise allow, the surplus food accumulating in the crop to be digested after the birds have gone to roost. The ability to store food has enabled some pigeons to be represented among the small list of birds—geese and certain galliform birds are other examples—that can live as efficient grazers, occupying a leaf-eating niche in wooded country, either in the tree canopy or on the forest floor. Some, such as the wood pigeon (*Columba palumbus*), have secondarily adapted to more open country to exploit the food supplies created by agricultural expansion. This pigeon is typical of the many that obtain their food partly in trees and partly on the ground. In midwinter it can subsist entirely on clover leaves and spends 95 percent of the day collecting the leaf fragments at 60–100 pecks per minute. By such intensive feeding, a single bird can collect 35,000 clover leaf fragments (45 grams, dry weight) per day. In spring this diet is supplemented with tree buds and flowers and also pasture-weed seeds; in summer the diet is chiefly cereal grain; in autumn declining corn stocks on the stubbles are supplemented with tree foods such as acorns and beech mast.

The mourning dove (*Zenaidura macroura*) of North America and the turtledove (*Streptopelia turtur*) and stock dove (*Columba oenas*) of Europe rarely take green vegetation, do not feed in trees, and so are examples of the trend toward complete ground feeding. These doves subsist almost entirely on seeds collected from low herbage or the ground. In winter such food sources become unproductive, so the turtledove and the mourning dove, in the northern part of their respective ranges, migrate southward. Turtledoves migrate from northern Europe to winter in central Africa. The ancients were well aware of this movement: "They shall hasten as a bird out of Egypt, and as a dove out of the land of Assyria" (Hosea 11:11).

Adaptation to ground feeding

Adaptation to ground feeding has progressed much further in various tropical genera, though the ecology of the species involved barely has been studied. In the New World a series can be traced from the typical *Zenaida* pigeons, through the ground doves (*Columbina, Claravis, Metriopelia* and *Scardafella*), which are mostly compact with short tails, and the *Leptotila* doves of South America, with longer legs suitable for running over the ground, to the New World quail doves (*Geotrygon* and *Starnoenas*). The last are relatively long-legged and partridgelike, spending most of their time on the forest floor. In the Australasian region there is a similar series, from scrub-country doves resembling *Zenaida* and *Streptopelia*—here represented by *Geopelia*—through the emerald doves and bronzewings (*Chalcophaps* and *Phaps*), to the Old World quail doves (*Gallicolumba* and various derivatives). Most of these quail doves live on the forest floor, collecting fallen seeds and fruits and seeds from low herbage, but in Australia a few are found in open country. Some aberrant species have evolved even further toward ground living, an extreme case being the pheasant pigeon (*Otidiphaps nobilis*) of New Guinea and adjacent islands, which resembles a pheasant. Evolution toward ground-living forms in Australasia and South America may have been helped by a paucity of mammalian carnivores that would normally be predatory on ground-living birds, which may partly explain why the extreme development of such forms has not occurred in Africa. In arid regions of Africa and central Asia, however, the various sand grouse (*Pterocles* and *Syrrphaptes*) occupy this niche.

The radiation of pigeons into fully arboreal, fruit-eating types also is most apparent in Australasia, with the brown fruit doves (*Phapitreron*), green pigeons (*Treron*), and the fruit doves (*Ptilinopus* and *Ducula*). Within these genera there are a large number of sympatric (coexisting) species differing by a wide variety of species-specific colour patterns that serve for mate recognition and hence species isolation. Nevertheless, the predominant colour of most fruit pigeons is green, which ensures effective camouflage when the birds are, as usually, among tree foliage. The success of fruit pigeons in this region may have resulted from the absence of competition from monkeys—important fruit eaters in Africa and South America. Fruit pigeons are adept at perching on slender branches and have large mouths that allow them to swallow large fruit. In tropical forests appropriate trees with ripe fruit tend to occur sporadically, so that most of the fruit pigeons move through large areas of country in flocks to seek out their food.

Sand grouse rely on cryptic coloration to protect them from predators yet are powerful fliers when disturbed. Long primary feathers facilitate considerable flights to water sources for some species, round trips of up to 80 miles being made daily in some areas. The desert species usually fly to water at dawn, those of bush country at dusk. At either time several thousand birds may congregate at a traditional watering place. Their drinking habits are unique. Only five to 10 seconds are required for drinking, an adaptation believed to reduce the risks from the diurnal predators that collect around waterholes in anticipation of the birds' arrival. Up to 15 millilitres (about 0.6 ounce) of water may be imbibed by a sand grouse. Water is carried back to nestlings in the crop and alimentary canal and on specially modified abdominal feathers. When not breeding, sand grouse are nomadic wanderers in their desert home, but some species undergo periodic irruptive dispersals that may involve long-distance movements, usually when good breeding leads to a high population and subsequent food scarcity.

Drinking habits

Behaviour. When feeding on the ground a pigeon must be looking downward for much of the time and therefore is vulnerable to predators such as foxes. The widespread habit of feeding in flocks has probably not evolved as an antipredator device but rather to help the birds to locate their food. Many species possess signal marks that become prominent in sudden flight, just as some rabbits display the white underparts of the tail. The wood pigeon has white wing bars that flash on takeoff, the Nicobar pigeon (*Caloenas nicobarica*) a white tip to the tail.

A consequence of feeding in a flock is that differences in individual attributes are more readily manifested, and this finds expression in a strongly developed social hierarchy (peck order). In competitive situations submissive individuals are frequently supplanted by more dominant individuals, and efforts to avoid conflict result in their getting less food. During times of shortage they are the first to die, other flock members remaining perfectly healthy. There is no evidence that social behaviour has evolved as a mechanism whereby populations can achieve self-regulation of their numbers, as has sometimes been suggested.

Reproductive behaviour

In most pigeons, a male in reproductive condition acquires a territory, which it proclaims with an advertisement call, usually a variant of the typical pigeon "coo" sound. Intrusion from other pigeons is prevented, at first by threat displays involving sleeking the plumage, stretching the head forward, and partially raising the wings. If these displays of aggressiveness are ineffective the male then attacks the intruder, pecking at the opponent and delivering sharp blows with the carpal joint of the wing. Females behaving submissively are gradually tolerated, particularly by unpaired males, enabling pair formation to proceed. In pigeons that build their own nests in scrub and open woodland, territories may be large, but in some hole- and cavity-nesting species, such as the rock dove, only the area around the actual nest site is defended. A shortage of suitable sites forces the latter to be semicolonial. Many pigeons perform display flights in and near the territory to attract unmated females. Such displays frequently involve exaggerated movements, slow wing beating, and, in some, loud clapping noises produced by the wings. The underparts of the tail, and sometimes other areas of the plumage, may have contrasting light and dark patterns that are displayed during flight.

Ronald K. Murton

Courtship behaviour.
One aspect of the wood pigeon's, or ringdove's (*Columba palumbus*), courtship is the prolonged, gentle rubbing of one bird's beak, usually the male's, through the head feathers of his mate.

Display flights

Pair formation has been well studied in the rock dove and Barbary dove. A sexually mature male typically approaches a submissive member of the same species with the bowing display (in most pigeons this occurs within a previously acquired territory, but not invariably). Bowing is primarily an aggressive display, involving tendencies to advance and attack and to mount and copulate. It is usually accompanied by vocalization—the bow call. Females normally retreat from the bowing display and exhibit submissive postures but if receptive sink to the ground with lowered wings and tail. This may result in the male mounting immediately; if not, he gradually becomes less aggressive and commences the nest-demon stration display. Being a submissive display, this reduces any female fear and attracts her closer, leading to a state where the pair now accept each other's presence. Displays that cement the pair bond involve mutual nibbling of each other's feathers, particularly those of the head and neck. At this early stage of the cycle the female exhibits juvenile behaviour, begging food from the male who, in turn, feeds her. This is the normal stage at which the female solicits and copulation occurs, interspersed with displacement preening. The female gradually assumes a more assertive role as she ceases to fear her partner and will attempt to push him off the nest site and occupy it herself. The male is now stimulated to search for nest material, which he brings back to the sitting female, and one or both incorporate the material into a nest.

Secretion of "pigeon milk"

The specific acts of seeing her mate build a nest and of being involved herself stimulate the female's neuro-endocrine system so that estrogen secretion rises, causing the oviduct to develop about five days before eggs will be laid. Seeing and incubating eggs, or seeing another bird doing so, stimulates prolactin secretion causing the crop gland to become secretory, so that "pigeon milk" is available when the eggs hatch. From this time onward the crop gland gradually regresses so that by the time they are ready to leave the nest, the young are fed almost entirely on environmental supplies. The production of a highly proteinaceous milk enables these seed- and fruit-eating birds to be independent of animal food when the young need maximum nutrition. The limitations of crop-milk production also have doubtlessly set the limit to the maximum number of young that can be reared, so that no pigeon ever lays more than two eggs, and many fruit eaters lay only one.

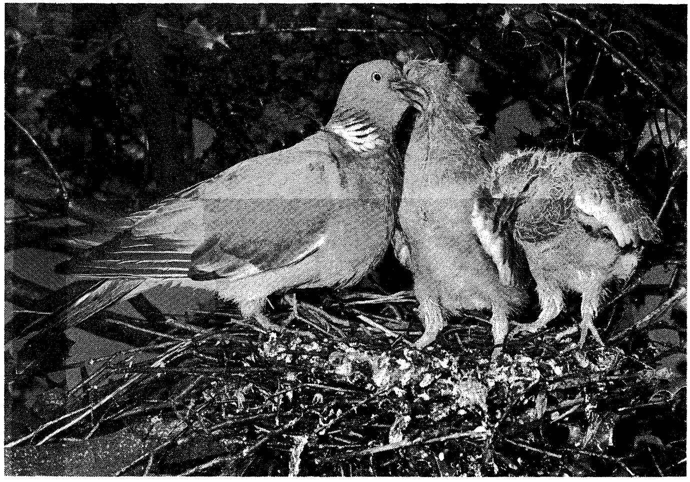
Most pigeons are multibrooded and have long breeding seasons, timed to correspond with the period of high food availability for both adults and young. Temperate zone pigeons, like other temperate zone birds, use proximate (immediate) environmental signals to insure attaining reproductive maturity at the appropriate time, one of the major signals being the seasonal changes of day length. The neuro-endocrine apparatus of British wood pigeons is stimulated by the day lengths of March; the gonads become active at that time and remain in breeding condition until September, when day lengths once more fall below the stimulatory threshold. This period of activity is appropriate to the season during which actual reproduction is feasible. The turtledove is stimulated into breeding condition in April during its migration north and lays its first eggs in late April and early May. The gonads spontaneously regress in August, inhibiting further activity, even though environmental conditions are good and breeding success is at a peak. The onset of reduced breeding activity in the presence of day lengths that were stimulatory in April is a safety mechanism; by reducing the physical strains associated with breeding, it enables the adults to accumulate the energy reserves needed to replace the flight feathers (vital for successful long-distance migration) and to lay down fat reserves for migration. Many pigeons, however, molt throughout the breeding season.

Nesting

Most pigeons build their own nests, simple flat platforms of twigs, in scrub or woodland. Some tree-nesting species occasionally nest on the ground, others regularly do so. White eggs are particularly attractive to predators so that, presumably to reduce these risks, some of the ground-nesting pigeons (*e.g.,* Old and New World quail doves) lay brown or cream-coloured eggs. In many pigeons, once the eggs are laid both members of the pair share the duties of incubation, changing over on or near the nest site, so that the eggs are normally not left uncovered. In contrast, sand grouse almost invariably lay three cryptically coloured and spotted eggs in a scrape on the ground; sometimes a few grass stems may serve as a lining. The male incubates by night, the female by day.

A few pigeons have adapted to nesting in crevices and holes, reducing the risks of egg predation but increasing the chances that the adult will be cornered on the nest by a predator. The clutch almost invariably consists of one or two eggs. The newly hatched young are at first continuously brooded but later are left for long periods while the parents seek food. The larger pigeons usually visit the nest only twice a day, bringing a crop full of food if feeding conditions are suitable. The young, called squabs, beg for food by pushing at the parent's breast, at the same time emitting a squeaky hunger note. They insert their bills in corners of the parent's mouth and are then fed by regurgitation. Although a pigeon is capable of rearing an artificial brood of three young, only two squabs can be fed at a time, and natural broods of three are extremely rare. In several species so far studied, when the nestlings are about three-quarters grown the parents begin a new cycle, building a new nest nearby and laying eggs.

Stephen Dalton—Natural History Photographic Agency



Feeding.
The unfledged squab of the wood pigeon (*Columba palumbus*) is nourished with "pigeon milk," a curdy substance regurgitated from the parents' crops.

Unguarded young are less subject to predation than are eggs. This is the case partly because of the squabs' more cryptic colouring, but primarily it is because they are capable of making a marked threat display. At the approach of danger they inflate their crops with air, causing them to become much bigger, and thrust their heads towards the enemy, at the same time hissing and snapping the mandibles together.

Ecology. Many temperate zone pigeons show marked population declines during the year. The high numbers resulting from breeding in temperate species can at first be maintained by extensive summer food supplies, especially in populations that depend on man-made food sources, such as barley or rice. But as food stocks decline in autumn and winter, reaching a critical level in relation to the population, juveniles suffer a high mortality, competing relatively unsuccessfully with adults. The change in population size from year to year thus depends primarily on the survival rate of juveniles through the period of food shortage, this being the key factor responsible for annual fluctuations. Long-term population trends, up or down, result from changes in the suitability of the habitat and are reflected in the number of adults that can settle to breed in an area. The average expectation of life for a British wood pigeon after reaching maturity is 2.25 years, which means that an average of 36 percent of the adult population dies each year. Whatever number of young 100 adults produce, only 36 need reach maturity to keep the population stable, and the rest are surplus. The juvenile mortality rate varies between 60 and 80 percent. Turtledoves are smaller and more at risk from predators; they also face the dangers of long migratory flights. Their adult mortality rate is nearer to 50 percent per annum, comparable to the 56 to 58 percent found in the North American mourning dove, which also is extensively shot for sport.

Population variability

FORM AND FUNCTION

Distinguishing characteristics. Pigeons are of compact shape, usually plump because of well-developed pectoral muscles, and with relatively small heads. The wings are long and often pointed in species that are highly migratory and in those that obtain most of their food in trees. A few island or montane species that fly less have reduced wings. A long, pointed tail, as in the extinct passenger pigeon and the masked dove (*Oena capensis*), is probably correlated with a high degree of manoeuvrability, necessary during a rapid escape from the ground in woodland. The partridge-like pigeons have short rounded wings and a short tail. These are mostly birds of woodland, keeping to the cover of trees and bushes, but in Australia there are species that live completely in the open and nest on the ground. One, the flock pigeon (*Phaps histrionica*), makes long flights to its feeding and drinking places and has long wings, in many respects apparently living like the sand grouse of Afro-Asian regions.

Pigeons generally have short legs, but in those that resemble game birds the legs are lengthened for more effective terrestrial locomotion. Three toes point forward and one backward. The bill is usually small and soft and may be overhung by the fleshy operculum (cere), which is enlarged in some of the fruit pigeons and domesticated forms of *Columba livia.* The bill shape is associated with feeding habits, slender bills being typical of seed eaters and deeply hooked bills of fruit eaters, especially those like *Treron,* which feed on large hard fruits such as figs. This trend in bill development is seen to an exaggerated extent in the ground-feeding tooth-billed pigeon (*Didunculus strigirostris*), whose bill resembles that of the extinct dodo, a bird that may have had similar feeding behaviour. It tears and nibbles its food into small pieces in a manner reminiscent of the parrots, taking berries, fruit, and mountain plantain.

Plumage

Pigeons have dense and soft plumage, the region in the vicinity of the eye often being bare. In most species the female is slightly duller than the male, but in some the sexes are identical, and in a few species there are marked differences in colour. One kind of sexual dimorphism, in which display plumage is confined to the male, is correlated in most other birds with a tendency toward polygamy; but it is not clear whether this is true in pigeons. Another kind of dimorphism involves the sexes' being rather differently coloured. Thus the male of the orange dove (*Ptilinopus victor*) is brilliant orange, the female green; the male ruddy quail dove (*Geotrygon montana*) is purplish chestnut, the female brown. This trend seems to be associated with making the female, who does most if not all the incubation in these cases, more cryptic.

With the exception of *Treron,* most fruit pigeons have a broad, short intestine and can void intact the stones from fruits they have eaten. Seedeaters have stronger gizzards and long, narrow intestines.

The sand grouse also are compact, with thick underdown, tough skins to resist desiccation, and frequently elongated central tail feathers. The legs are short and the toes sometimes slightly webbed, aiding progression on soft sand.

Physiology and biochemistry. Domesticated pigeons and the Barbary dove have long served as subjects for avian physiological research, and knowledge related to their body functions is extensive. They appear to exhibit no remark-

Production of prolactin

able specializations, compared with birds in general, with the exception that the crop becomes glandular in response to small amounts of the hormone prolactin. Prolactin was first discovered in pigeons by O. Riddle, and pigeons still serve in the bioassay of this hormone from other sources. Prolactin also is produced by other birds. Among other functions, it reduces aggressive behaviour during the incubation and early brood-care stages of the reproductive cycle. It apparently is involved in the molting process and in mechanisms associated with preparation for migration. Pigeons are unique only in having secondarily evolved a new target organ responsive to the hormone.

The sense of taste, as with most other birds, is poorly developed, and it is probably not an important factor in selecting food. Shape and tactile characteristics and, to a lesser extent, colour are much more important. The average number of taste buds is only 37, confined to the soft area at the base of the tongue and palatine region (in contrast, man has about 9,000 and the rabbit about 17,000). Pigeons can, however, exhibit a surprising sensitivity to certain substances, such as acids; only at extremely low concentrations are acetic acid solutions accepted as readily as pure water. The olfactory organs are well developed; but smell seems to be of little significance in the daily lives of pigeons, and experiments to demonstrate their olfactory abilities have yielded conflicting results. Visual acuity is highly developed, as in most other birds, although training experiments demonstrate that pigeons can attain an acuity little better than humans.

EVOLUTION AND PALEONTOLOGY

Dodoes and solitaires

By the time they appear in the fossil record the Columbiformes are already so well differentiated that their phylogenetic relations cannot be determined with certainty. The sand grouse and pigeons resemble each other anatomically, but this may have resulted from convergence towards a similar mode of life. The dodoes and solitaires were highly specialized island forms that doubtless arose in the Mascarene Islands and were peculiar to those islands. Three species are known, the dodo (*Raphus cucullatus*) on Mauritius, the Réunion solitaire (*Raphus solitarius*), and the Rodrigues solitaire (*Pezophaps solitaria*). A fourth species, the white dodo (named as *Victoriornis imperialis*), may have existed on Réunion but, if it existed, was probably a colour phase of the Réunion solitaire. The dodoes and solitaires became extinct in about 1681, 1746, and 1791 respectively, as they fell easy prey to marauding sailors and could not compete with pigs and other introduced livestock. They were pigeon-like birds that had lost the power of flight in the safety of their predator-free island existence and had become large (as big as a turkey), heavily built birds with strong bills and feet. The wings had become rudimentary, and the sternum possessed only a small keel.

The earliest known sand grouse are of the modern genus *Pterocles* and date from the upper Eocene of France (about 40,000,000 years ago). The earliest known pigeon is *Gerandia calcaria* from the early Miocene of France (25,000,000 years ago), although the suborder probably arose in the Australasian region, where the greatest variety of modern columbiforms is found. Other fossil pigeons are known from the Pleistocene (2,000,000 years ago).

CLASSIFICATION

Distinguishing taxonomic features. The most important features distinguishing the Columbiformes are the structure of the skull, sternum, and furculum. The several species have close thick feathers set loosely in the skin, lack the fifth secondary feather of the wing, and possess a crop. The structure of the bill and nostrils, the arrangement of the sterno-tracheal musculature, and the presence or absence of intestinal ceca distinguish the suborders. Behavioral characteristics are of doubtful validity in relating suborder relationships because the similarities may have arisen by convergence. Blood-antigen relationships may be considered more reliable.

Annotated classification.

ORDER COLUMBIFORMES

Birds in which the palatine processes of the maxillae do not meet in the midline (schizognathous) and with vomer absent or vestigial. Palatine and pterygoid bones articulate with basisphenoid rostrum. Basipterygoid processes present, except in family Raphidae. Sternum with large lateral and smaller inner incisions, these often fused to fenestrae. Furculum U-shaped and with hypocleideum little or not developed. Nares separated by a complete internasal septum. Hypotarsus complex. Plumage, close and thick, feathers loosely set in skin. Pterylosis essentially similar in the 2 suborders. 5th secondary quill absent (diastataxic). Oil gland naked or absent. Well-developed crop present.

Suborder Columbae (pigeons, doves, and dodoes)

Miocene to present. Intestinal ceca absent or minute, syrinx with sterno-tracheal muscles asymmetrical. Young altricial (helpless); hatched blind without real down. Basal part of bill soft and covered with a swollen skin that envelops the slitlike nostrils as a cap (operculum). External nasal opening into skull tapers behind into a narrow cleft running back into nasal bone (schizorhinal); 14–15 cervical vertebrae. 11 primary feathers, the outermost much reduced. Usually 12–14, and exceptionally 16–20, retrices. Crop bilobed. Nest in tree holes or caves or build flimsy platform of sticks on trees or ground. Eggs usually white, except in some quail doves brown or cream, mostly 2 in number, but 1 in some genera.

Family Columbidae (pigeons and doves)

Pleistocene to present. Characteristics of suborder. Length 15–80 cm; weight 45–4000 g. Wide range of colours, from grays and browns to striking orange, green, or purple. Worldwide except subpolar regions and some oceanic islands. About 285 species.

Family Raphidae (dodoes and solitaires)

Extinct but with no fossil record. Flightless, with much reduced furculum and wing, fused coracoid and scapula and no basipterygoid processes. Large; weight probably exceeded 10 kg. Limited to Mascarene Islands. 3 species.

Suborder Pterocletes.

Long intestinal ceca, symmetrical sterno-tracheal muscles, young nidifugous (precocious) and covered with down. Bill resembles that of game birds, not soft at base. Sternum with very high keel and ribs broad, cervical vertebrae number 15 to 16; tarsus short and feathered, toes wide and short, soles covered with hexagonal or rounded scales. Nostrils not schizorhinal and covered by feathers except in front. 11 primaries, outermost much reduced, and 14–18 retrices. Pterylosis as in pigeons but feather tracts wider. Oil gland small and naked. Large crop, not bilobed. 2 to 4 eggs laid in mere scrape in ground in open steppe or desert.

Family Pteroclididae (sand grouse)

Late Eocene or early Oligocene to present. Characteristics of the suborder. Coloration buff, reddish or grayish brown, barred or spotted with black, brown, yellow, or white; sexually dimorphic. Habitat sandy plains, open brush, or deserts of temperate to equatorial Eurasia and Africa. Length 23–40 cm (variation mostly in tail length); weight 120–300 g. 16 species.

Critical appraisal. On anatomical grounds, the sand grouse resemble the pigeons and are therefore placed in the same order. Their drinking behaviour only partly resembles that of the pigeons, and other behaviour in which the two groups are similar is not unique to either. G.C. McLean suggested that the sand grouse are nearer to the plovers (family Charadriidae, order Charadriiformes), and D. Goodwin also took this view. The subordinal limits of the Columbae seem well defined, although J. Lüttschwager suggested that the Raphidae bear closer affinity with rails (order Gruiformes). The gruiform line diverged early from the stock that eventually produced the Columbiformes and Charadriiformes, so if closer to the Gruiformes the dodoes would be closer to presumed pigeon ancestors, rather than, as usually believed, relatively recent offshoots from established Columbae stock. The pigeons and doves comprise a natural group, which most authors treat as a single family. This seems the best arrangement, but separate family rank is sometimes given to the pheasant pigeon, crowned pigeons, tooth-billed pigeons, and even to the fruit pigeons. These are probably best treated as the subfamilies Otidiphapinae, Gourinae, Didunculinae, and Treroninae. (R.K.M.)

Psittaciformes (parrots, lories, cockatoos)

The avian order Psittaciformes contains more than 300 species of generally brightly coloured, noisy, tropical birds, Early association with man

to which the general name parrot may be applied. Various species are known as keas, cockatoos, cockatiels, lories, lorikeets, parrotlets (or parrolets), parakeets, budgerigars, rosellas, conures, lovebirds, amazons, and macaws. Although the first accurate written reference to a parrot is frequently credited to the Greek historian Ctesias, of the 5th century BC, who described clearly what is now called the blossom-headed parakeet (*Psittacula cyanocephala*) of India, there is no doubt that parrots were associated with man much earlier, for aborigines on all continents had parrots as pets when first visited by explorers. Parrots of many kinds have been long transported to Europe for zoos and private collections. Affluent citizens of early Rome often kept parrots in their homes and even esteemed them as delicacies of the dinner table.

GENERAL FEATURES

Bodily proportions. Parrots vary in total length from eight to 100 centimetres (three to nearly 40 inches), the latter in long-tailed forms. The short neck and sturdy body, along with the stout feet and thick bill, give them a bulky appearance. The broad wings are often pointed; the tail is highly variable in both length and shape. In some species the tail is short and rounded or square; in others, such as the macaws (*Ara*), it is long and pointed. In numerous species the central tail feathers are very long, surpassing the body in total length. In the five species of racket-tailed parrots (*Prioniturus*) the central tail feathers are longer than the others and are spatulate, the middle part of the feather shaft being bare. No parrot has a forked tail. Pointed wings and a long tail usually are found in species that fly great distances; rounded wings and blunt tails typify the more adept climbers. Most parrots are swift on the wing, although they normally fatigue quickly.

Extinction of Carolina parakeet

Distribution. Parrots are primarily birds of the tropics. Their distribution encompasses the tropical and south temperate regions of the world, including Madagascar, New Zealand, and the West Indies. In Asia they occur throughout almost all of India but extend northward only to the Himalayas and southern China. They are absent from Europe. In the early 1970s in North America they ranged north only to the extreme southwestern United States (one species, the thick-billed parrot, *Rhynchopsitta pachyrhyncha*). Prior to the early 1900s, however, the Carolina parakeet (*Conuropsis carolinensis*) inhabited most of the eastern United States; it was brought to extinction by human persecution. The last captive died in the Cincinnati Zoological Garden in 1914, but the last generally accepted observation in the wild was a flock seen in Florida in 1920, although it has been claimed that they existed in South Carolina until 1938. In the Southern Hemisphere a number of parrots range to Tasmania and New Zealand, and in South America one species is found on Tierra del Fuego, but none are found on the southern tip of Africa.

Although parrots are found in most parts of the three southern continents (excluding the Sahara), they are not evenly distributed. Of the 81 genera recognized in a 1937 revision of the taxonomy of parrots, the Neotropical region (South and Central America) has 28, not one of which is found elsewhere. The Nearctic region (temperate and arctic North America) has had only the now-extinct genus *Conuropsis.* The Australo-Papuan region, which encompasses Australia, New Zealand, New Guinea, Celebes, and many of the Pacific Ocean islands, has 44 genera, three of which are shared with the Oriental region. The Oriental region, which, in addition to Asia, includes the Philippines and most of Malaysia, has nine genera, including the three shared with the Australo-Papuan region and one shared with the Ethiopian region. The Ethiopian region, comprising Africa, Madagascar, and nearby islands, has only six genera, one of which is the extinct *Mascarinus* from Reunion Island, 650 kilometres (400 miles) east of Madagascar. The Palearctic region (Europe and northern Asia) has none. Most parrot genera contain only a few species; only six include more than 10 species, four of which occur in the Neotropics.

Parrots as pets

Importance to man. The qualities of parrots, especially the ability of many species to imitate human sounds, make them popular as pets. The African gray parrot (*Psittacus erithacus*) and some species of amazons (*Amazona*) from the New World tropics are particularly good mimics. There is no evidence to suggest, however, that talking parrots realize what they are saying. Another appealing attribute of parrots is their display of affection, not only to others of their own species but also to man. Pairs of many species, especially the lovebirds (*Agapornis*), are together almost constantly, nibbling each other's feathers with seeming affection; if one bird disappears its mate sometimes dies, apparently of loneliness. Many parrots seem to delight in being petted and scratched, which is rare among birds. Parrots have extremely powerful jaws, however, and an indiscriminate attempt to pet them can result in a severe bite. The use of the toes for climbing and food handling, in much the same manner as man uses his hands, also makes parrots appealing. Their longevity, bright colours, intent gaze, ability to learn tricks, and willingness to remain on a perch instead of fluttering about contribute to man's fondness for various kinds of parrots as pets. Finally, most species are vegetarians and thrive on a varied diet. This circumstance, as well as the fact that their droppings typically are dry and compact, means that parrots require little care.

Most species of parrots have been kept in captivity at one time or another, and most have been bred, with the exception of the pygmy parrots (*Micropsitta*). A large zoological garden may have more than 100 species on display at one time. No parrot has been domesticated in the sense of gallinaceous birds and waterfowl, although in recent decades breeders have produced a variety of

Drawing by G. Tudor

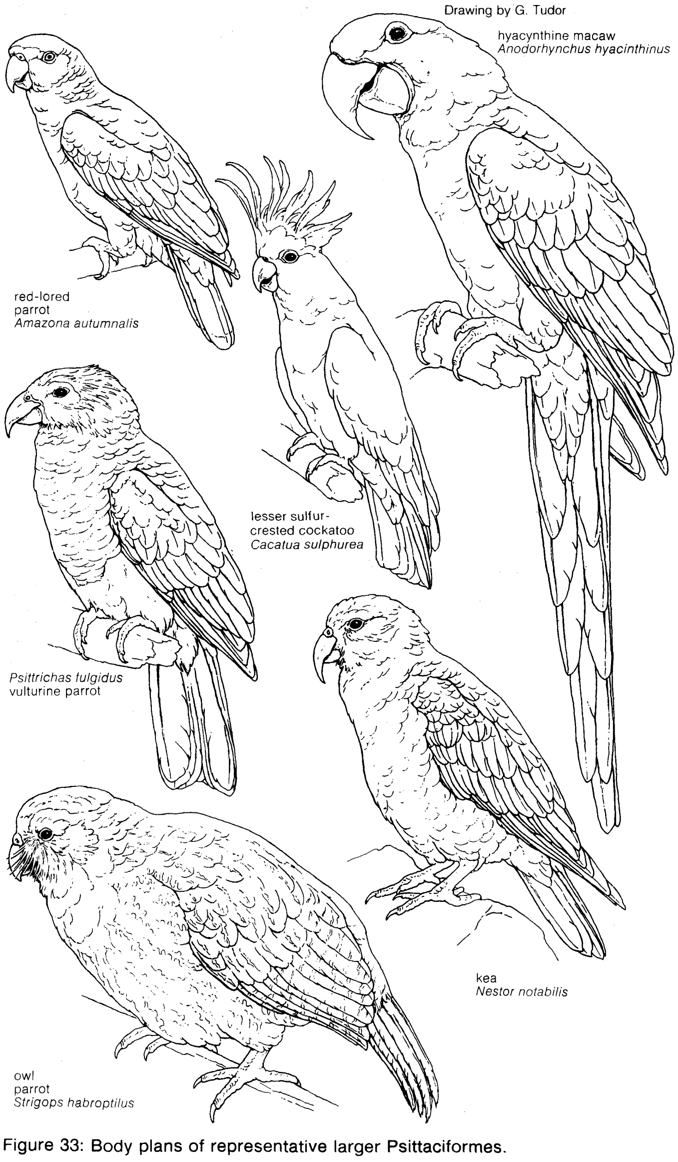


Figure 33: Body plans of representative larger Psittaciformes.

Longevity in parrots

colour strains of the budgerigar (*Melopsittacus undulatus*), commonly called shell parakeet. In the mid-1950s "budgies" became popular household pets in the United States; within a decade more than 5,500,000 people had at least one in their homes. Captive parrots, especially the larger species, are long-lived. Claims of 80 or even 100 years are frequent and perhaps true, although thus far impossible to document. A convincing record of a 56-year-old greater sulfur-crested cockatoo (*Cacatua galerita*) exists.

In the early 1930s the importation and sale of parrots in the United States was drastically curtailed by quarantine laws designed to combat psittacosis, or parrot fever (better called ornithosis, as the disease by no means is restricted to psittacine birds), a respiratory virus that can infect humans. Antibiotics now available reduce the severity of the disease, and most of the restrictions on the parrot trade have been relaxed.

The primary economic importance of parrots derives from their popularity with aviculturists. So popular are some species that the Australian government has found it necessary to pass laws forbidding export of the birds because wild populations were being decimated. Unfortunately, only a few of the many individuals captured ever reach the comparative safety of a comfortable cage, because of mishaps en route. Some parrots, especially certain of the Australian seed-eating species, damage crops and therefore are hunted and killed.

NATURAL HISTORY

Habitat and food choice. Most parrots inhabit forests, although a few live in grasslands. Of the forest-inhabiting species many forage along the forest edge and on the ground. Some parrots live in mountains, especially in the Himalayas and Andes; the New Zealand kea (*Nestor notabilis*) is a mountain inhabitant but obtains much of its food in forested valleys; it nests either in high elevation forests or near the forest edge. Many Australian parrots, such as members of the genera *Neophema* (grass parakeets) and *Psephotus,* are found in dry, open grasslands, typically where trees are scattered through the habitat. The budgerigar and the rare night parrot (*Geopsittacus occidentalis*) also are Australian grassland birds.

Drawing by G. Tudor

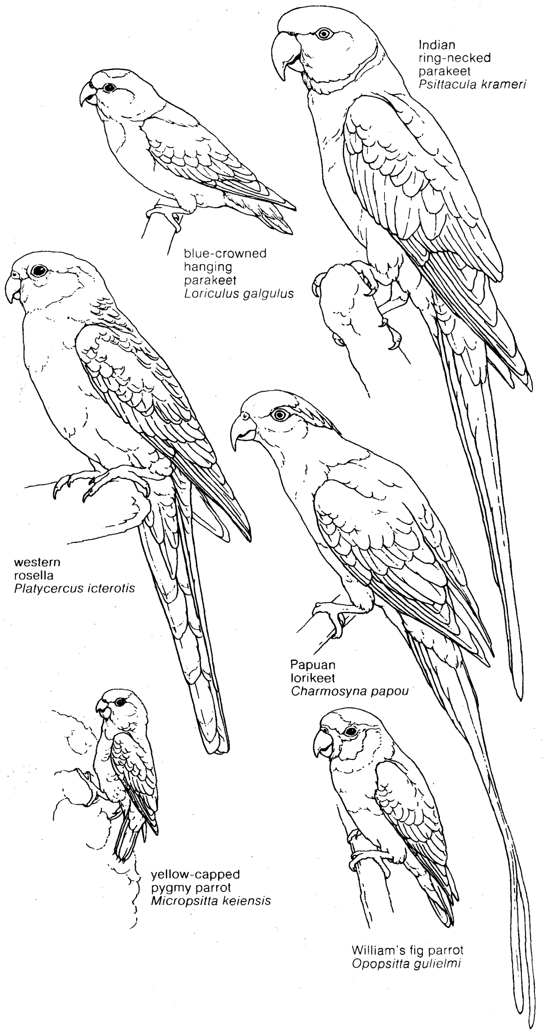


Figure 34: Body plans of representative smaller Psittaciformes.

Parrots feed almost entirely on plant materials. The smaller species tend to utilize grass seeds, berries, fruits, and the juices of blossoms; the larger forms obtain fruits and nuts from trees, and bulbs, tubers, and roots from the ground. When digging, many parrots also capture larval and adult insects, and raven, or black, cockatoos (*Calyptorhynchus*) gnaw through bark to obtain wood-boring beetles. Many kinds of nectar-eating birds suck juices through tubelike tongues, but brush-tongued parrots feed on nectar by crushing flowers and licking the juices. The tiny pygmy, or woodpecker, parrots feed on fruit, arboreal termites, and fungi. The kea feeds on dead sheep and carrion and will attack sick, injured, or trapped individuals but rarely will it harm healthy sheep.

Aggregations of parrots

Social behaviour. Typically, parrots are gregarious and noisy, often forming small groups—sometimes huge flocks—flying rapidly high overhead and screeching. Their seemingly conspicuous bright colours are somewhat misleading, for a group of parrots in foliage is difficult to discern. The grassland-inhabiting parrots are nomadic and often occur in flocks of tens or even hundreds of thousands. The development of agriculture in the interior of Australia, particularly the increased availability of water, has resulted in larger populations of several species, such as the corella (*Cacatua sanguinea*) and budgerigar.

The vocalizations of most parrots are loud, raucous screeches; generally the larger the species the more ear-splitting the calls. The voices of some of the smaller ones include chattering and twittering notes pleasant to the human ear. About 12 different calls, each announcing a different mood, have been identified for the greater sulfur-crested cockatoo. The amazing mimetic abilities of many parrots mentioned above are expressed only in captivity.

Reproduction. Parrots are monogamous. Some species breed colonially; others space themselves through the nesting habitat. Based on the fewer than 10 species that have been studied extensively, courtship and behaviour to maintain the pair bond may include vocalizations, bill-caressing, mutual preening, bowing, wing-raising, tail-spreading, and feeding the mate.

Nesting habits

With few exceptions, parrots nest in holes in trees. Some species add nest material such as leaves, fibres, and bark strips; others lay their eggs on the floor of the cavity. Some lovebirds cut leaves into strips, which are then tucked into the feathers of the back for transportation to the nest. Several parrots, including the pygmy parrots and the orange-fronted parakeet (*Aratinga canicularis*), hollow out cavities in termite nests. Exceptional in the family is the monk parakeet (*Myiopsitta monachus*) of South America, which builds a communal stick nest in trees. Several species nest in rock crevices or earthen caves; examples include the burrowing parrot (*Cyanoliseus patagonus*), kea, night parrot, and the flightless owl parrot (*Strigops habroptilus*). The ground parrot (*Pezoporus wallicus*) lays its eggs in a shallow cup on the ground.

Parrot eggs are white and usually nearly spherical. Generally the larger species lay only two eggs and produce one brood per year; the smaller species lay up to six or even eight or nine eggs and may breed two or three times per year. Incubation time, which generally varies directly with size, ranges from 16 to 30 days in the few species for which there is documentation. Either both parents or else only the female may incubate the eggs. The young hatch naked or with very sparse down. They are altricial—that is, they are helpless and require complete parental care—and they are also nidicolous—that is, they remain in the nest for some time after hatching. The young are fed by regurgitation, typically by both parents. In some species at least, care of the young may continue for several weeks after they have left the nest.

FORM AND FUNCTION

The yoke-toed foot

Foot structure. Parrots can be distinguished from other birds by the structure of the feet and bill. Most birds have the four toes arranged with three directed forward—the inner (II), middle (III), and outer (IV)—and one backward, the hallux (I). This condition, called anisodactyl, literally means without equal toes, referring to the unequal arrangement. Parrots have two toes (the inner and middle) directed forward, and two directed backward; this arrangement is called zygodactyl, which literally means yoke-toed and refers to the occurrence of toes in pairs. Zygodactyly also occurs in woodpeckers and their allies (Piciformes), cuckoos (Cuculiformes), and some other birds. The proximal (upper) bone of a bird's foot, the tarsometatarsus (commonly considered the lower leg), lies between the elevated heel joint and the toes. In parrots it is short and stout, and at least one toe is always longer. It is the characteristic short, thick tarsometatarsus—or tarsus, as the entire region is called—and the zygodactylous, long, strong toes that enable parrots to climb and manipulate objects so ably. The entire foot is encased in tough skin covered with small scales.

Bill and skull. The most distinctive morphological trait of parrots is the strongly hooked, powerful bill. Superficially the bill resembles that of the hawks and owls, but the upper and lower mandibles of parrots normally have a stronger and more uniform curve. Often the anterior edge of the lower jaw is broad and truncate. The under surface of the upper bill usually possesses transverse or oblique file-like corrugations where the lower jaw occludes. These file-like ridges, along with the highly manipulative tongue, assist in holding seeds as the bird uses the chisel-shaped cutting edge of the lower bill to peel away a seed cover.

The use of the bill for manipulating objects, cracking hard nuts, and as a third "foot" in climbing are all possible because of a highly kinetic (movable) upper jaw. Most living birds have such a kinetic upper jaw, which is connected to the skull dorsally by a hinge, and is able to be moved independently by swinging on this hinge, but nowhere among birds is this kinesis better expressed than in parrots. The raising of the upper jaw can be described as follows: all movement of the upper jaw originates at the point of attachment between the skull and the quadrate bone, which forms the hinge between the skull and the lower jaw. Two series of bones (the quadratojugal-jugal series and the pterygoid-palatine series), both of which lie in the roof of the mouth, are situated between the quadrate and the upper jaw. When the quadrate is swung forward, the two series of bones slide forward, causing the upper jaw to swing upward on its hinge with the skull.

The short, thick, and fleshy tongue of parrots shows a variety of specializations at the tip; one found in several groups is a brushlike fringe. Primarily, the tongue functions to manipulate and hold food. Those parrots with brushlike terminal papillae (nipple-like projections) on the tongue use them to hold juices, as a brush holds paint.

All parrots possess a cere, an area of soft skin surrounding the nostrils; it may be naked or covered with small, soft feathers. In adult budgerigars the cere is blue in males and tan in females.

The orbits (eye sockets) of some, but not all, parrots are ringed with bone. Other features of the parrot skeleton include a prominent keel on the sternum (breastbone), except in the flightless owl parrot of New Zealand, and a highly variable furcula (wishbone), which may be normal, weak, unfused medially, or absent.

Plumage coloration

Skin and plumage. Parrots have relatively few feathers, which are hard in texture and normally gaudy in colour. Many species are bright green with patches of red, orange, yellow, blue, or white; the plumage of others is predominated by the latter colours. A few parrots are brown or all green. Sexes are alike or nearly so, with a few notable exceptions. One, the eclectus parrot (*Lorius roratus*), was for many years thought to be two separate species until it was noted that only males were known for the predominantly green "species" and only females for the wine-red "species." The head is crested in a few parrots, especially among the cockatoos (Cacatuinae).

Powder downs, which occur in a variety of birds including some parrots, are specialized feathers, usually found in well-defined patches, that produce a powdery substance used to clean and waterproof the other feathers. They are well developed in cockatoos, in which they occur primarily as a pair of lateral rump patches.

Skin glands, which are abundant in mammals, are almost entirely lacking in birds, with the exception of the oil gland. The oil produced in this gland—also known as the uropygial gland because of its location at the base of the tail and as the preen gland because of its function—is used, like the powder down, to clean and waterproof the feathers. Oil is squeezed from the gland, and birds either use the bill to apply it to the feathers or rub their heads directly over the gland. The nipple of the gland, which protrudes through the skin at the base of the tail dorsally, is surrounded by a tuft of feathers in parrots. Not all parrots have an oil gland; for example, it is absent from the *Amazona, Brotogeris,* and *Pionus* parrots and greatly reduced in some others.

EVOLUTION AND CLASSIFICATION

The extinction of island species

Evolution. Dispersed on a multitude of small islands, parrots have always been vulnerable to extinction, and in recent times the group has suffered increasingly in this regard. From 1680 to the early 1960s at least 16 species disappeared entirely, and another 14 became endangered. Most of the extinct species lived on small islands or on large islands in the West Indies; only the Carolina parakeet had an extensive continental range.

The place of evolutionary origin of the parrot family is not known. The greatest structural diversity is exhibited by parrots from the Australo-Papuan region, but the greatest number of species occur in the Neotropical region. Some authorities doubt that either was the original home of this group, although it probably did arise in the Old World. The scanty fossil record is of no assistance in determining the original home of parrots, but it does provide an indication that the group is very old. Typical parrots are known from fossils from the Miocene epoch (about 20,000,000 years ago) of France and from North America. The fact that the Old and New World share no genera and that five of the six Ethiopian genera are not found elsewhere supports the antiquity of the group. An American biologist, Philip J. Darlington, has suggested that parrots and pigeons (Columbiformes), which show similar characteristics of distribution, may have been dominant everywhere before the rise of perching birds (Passeriformes), which may now be replacing them.

Annotated classification. The following classification follows the scheme proposed in 1959 by the German ornithologist Hans von Boetticher. He treated the parrots, as did most of his predecessors, as one family, with seven subfamilies, the two largest of which are further subdivided into tribes. Boetticher reduced the 81 genera recognized by U.S. ornithologist James Lee Peters in 1937 to 75, a figure that reflects the modern view of broader genera.

ORDER PSITTACIFORMES (parrots and allies)

Chunky, primarily tropical birds, with short necks, tarsi, and wings. Distinctive bill, short and strongly hooked, the upper mandible extending down over the tip of the upcurved lower mandible. Feet zygodactyl. Most brightly coloured; often gregarious; highly vocal. About 75 genera, 310–320 species; length 8 to 100 cm; found in the tropics and subtropics of the world and the temperate regions of the Southern Hemisphere.

Family Psittacidae

Characteristics of the order.

Subfamily Nestorinae (kea and kaka)

Bill rather long, narrow, less curved than in other parrots. Plumage greenish-brown. Two species: kea (*Nestor notabilis*) and kaka (*N. meridionalis*); length about 45 cm; New Zealand and adjacent islands.

Subfamily Psittrichasinae (vulturine parrot, or bristlehead)

Bill rather elongated, but more curved than in Nestorinae; face largely devoid of feathers, those present stiff and narrow. Plumage dark brown to black, with large patches of red. One species, *Psittrichas fulgidus;* length about 45 cm; New Guinea and adjacent islands.

Subfamily Cacatuinae (cockatoos)

Bill strongly curved (and massive in the palm cockatoo, *Pro-*

bosciger aterrimus); lower mandible wider than upper. Plumage black, gray, pink, or white, sometimes tinged with yellow or pink; often a prominent erectile crest, sometimes yellow or red. Four (or 5) genera, about 16 species; length about 30–80 cm; Australo-Papuan region and Philippine Islands.

Subfamily Micropsittinae (pygmy parrots)

Tiny parrots. Tail feathers short, with stiff shafts and pointed tips. One genus, 6 species; length about 8 cm; New Guinea and adjacent islands.

Subfamily Trichoglossinae (lories and lorikeets)

Bill relatively small; tongue brush-tipped. Mostly brightly coloured and gregarious. Feed on fruit, buds, pollen, and nectar; may be important pollinators of trees. Two tribes, Psittaculirastrini and Trichoglossini, about 14 genera, 62 species; length 11–32 cm; Australo-Papuan, Polynesia, Indonesia, and the Philippines.

Subfamily Strigopinae (owl parrot or kakapo)

Large, flightless parrot with soft greenish plumage and an owl-like facial disk. Nocturnal and vegetarian. One species, *Strigops habroptilus;* length about 60 cm; New Zealand.

Subfamily Psittacinae (all other parrots)

Range of characteristics encompasses that of order. About 52 genera, 234 species.

Tribe Platycercini (rosellas and allies). Most with long tails, many with bright plumage colours. Twelve genera, 31 species; length about 17–40 cm; Australo-Papuan, especially in interior of Australia. Tribe includes budgerigar (*Melopsittacus*), grass parakeets (*Neophema*), rosellas (*Platycercus*), night parrot (*Geopsittacus*), and ground parrot (*Pezoporus*).

Tribe Loriini (wax-billed parrots). Bill typically smooth and waxlike. Many species with bright plumage, sexes often differing. About 14 genera, 47 species; length about 14–45 cm; Australia and southern Asia to Africa and Madagascar. Tribe includes racket-tailed parrots (*Prioniturus*), ring-necked parakeets (*Psittacula*), the eclectus parrot (*Lorius*), and lovebirds (*Agapornis*).

Tribe Loriculini (bat parrolets or hanging parakeets). Small, short-tailed, green parrots. Rest hanging upside down. One genus (*Loriculus*), 9 species; length about 12 cm; India to Philippines and New Guinea.

Tribe Psittacini (blunt-tailed parrots). Medium-sized parrots; many brightly coloured, especially in green. Many are good mimics, especially in genera *Psittacus* and *Amazona.* Twelve genera, 66 species; length about 16–65 cm; New World tropics (majority), Africa and Madagascar.

Tribe Araini (wedge-tailed parrots). Tail usually graduated (*i.e.,* centre feathers longest, side feathers progressively shorter), often long and pointed. Some, especially the macaws (*Ara*) and blue macaws (*Anodorhynchus*), brilliantly coloured in red, yellow, and green, or blue. Many species with bare skin around eyes. Thirteen genera, 81 species; length about 12–100 cm; New World tropics. Three genera collectively include 54 species: *Pyrrhura* and *Aratinga* (conures and parakeets) and *Ara* (macaws).

Critical appraisal. With the present paucity of knowledge about the evolution of the orders of birds, any asserted or supposed relationships of the Psittaciformes, which are highly specialized, are necessarily purely hypothetical. Various authors have suggested, with little supporting evidence, relationships to such groups as the hawks (Falconiformes), owls (Strigiformes), game birds (Galliformes), pigeons (Columbiformes), woodpeckers (Piciformes), and cuckoos (Cuculiformes), especially the turacos (Musophagidae).

Relationships within the parrot order have proved just as perplexing. In the 1890s British ornithologist Alfred Newton commented on the state of knowledge of parrot classification in *A Dictionary of Birds* by writing:

> It is a reproach to ornithologists that so little satisfactory progress has been made in this direction, and the result is all the more disheartening seeing that there is no group of exotic birds that affords equal opportunities for anatomical examination, since almost every genus extant, and more than two-thirds of the species, have within recent times been kept in confinement in one or another of our zoological gardens, and at their death have furnished subjects for dissection.

Since Newton's time at least six classifications of the order have been proposed, many of which show basic similarities, but his statement was still basically true in the late 20th century. (G.E.W.)

Cuculiformes (turacos, cuckoos, roadrunners)

The bird order Cuculiformes is a cosmopolitan group containing two very distinct families, the cuckoos, Cuculidae, and the turacos, or plantain eaters, Musophagidae. Discussion of their relationship has led authorities increasingly to give the turacos ordinal rank. The family Cuculidae is much the larger group, containing about 127 species, found in the tropical and temperate zones of all the continents except Antarctica and on many oceanic islands; the Musophagidae contains 20 species, found only in Africa.

The cuckoos are an ancient group with uncertain phylogenetic affiliations and no living near relatives, even the turacos being quite distinct and with no intermediate, or connecting, species. The cuckoos are of unusual biological, especially ethological, interest because many species are brood parasites; *i.e.,* they lay their eggs in the nests of

Drawing by R. Keane

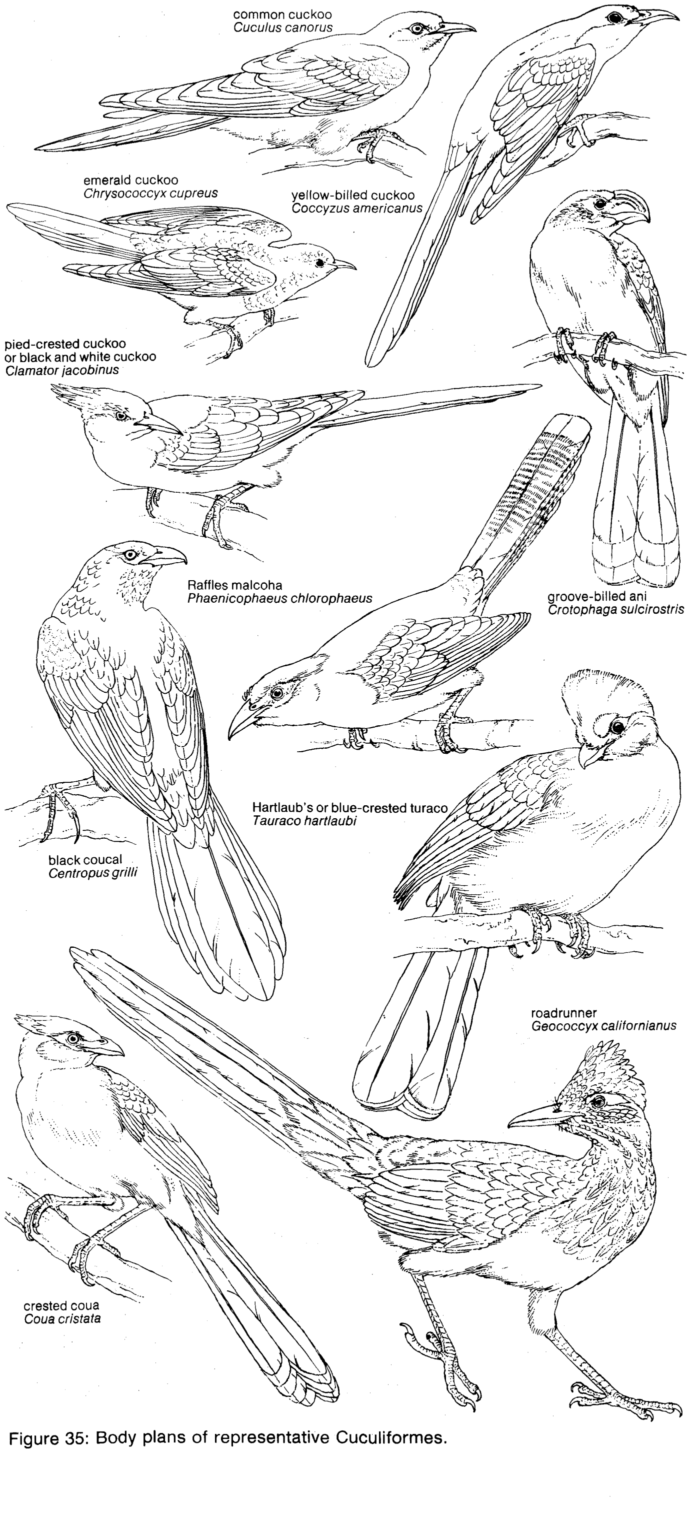


Figure 35: Body plans of representative Cuculiformes.

other species, which then rear the young cuckoos. Other cuckoos make their own nests, in which they incubate their eggs and rear their young as do most birds; and still others (*Crotophaga* and *Guira*) build communal nests. Some cuckoos are among the few birds that feed extensively on hairy caterpillars.

GENERAL FEATURES

Size ranges

The cuckoos cover a great range in size, from the small glossy or emerald cuckoos of the genus *Chrysococcyx,* which are about 15 centimetres (six inches) long, to the large ground cuckoos (*Carpococcyx*) and the larger species of coucals (*Centropus*), which reach nearly 90 centimetres (three feet), including the tail, which is often strikingly long. Most cuckoos have fairly loose-webbed feathers, varying in colour from subdued browns, grays, olive, and black, to brilliant, iridescent greens and purples and bright yellow. The beak is of moderate length and often slightly downcurved. The turacos are all sizable birds, the smallest species having a total length of about 37 centimetres (15 inches) and the largest attaining a length of almost 75 centimetres (30 inches).

Structural diversity in cuckoos

In keeping with the greater number of their included species and their worldwide distribution, the cuckoos show far more diversity in structure than do the relatively homogeneous turacos. The cuckoos include the arboreal "typical" cuckoos of both the Old World and the New World, the terrestrial roadrunners (*Geococcyx*) of southwestern United States and Mexico, and the more compact but also largely terrestrial coucals (*Centropus*) of Africa and Australasia.

DISTRIBUTION

Although the family Cuculidae is virtually worldwide in temperate and tropical regions, most of the subfamilies are restricted to one hemisphere or the other. Three (Cuculinae, Centropinae, and Couinae) are limited in distribution to the Old World; the subfamily Crotophaginae is wholly New World; the Neomorphinae is mostly New World, with one genus in Southeast Asia; and the sixth subfamily, Phoenicophaeinae, is represented in the tropics of both hemispheres. Many cuckoo genera are peculiar to certain parts of the world; others are cosmopolitan. Three genera (*Rhamphomantis, Caliechthrus,* and *Microdynamis*), for example, are known only from New Guinea; *Dasylophus* and *Lepidogrammus* occur only in the Philippine Islands; and *Saurothera* (lizard cuckoos) are found only in the West Indies. Some other genera are wide ranging: the typical cuckoos of the genus *Cuculus* are found in Europe, Asia, Africa, and Australia; the small, brightly coloured *Chrysococcyx* species occur in Africa, Asia, Australia, and some Pacific islands; and the coucals (*Centropus*) live in Africa, Asia, Australia, and adjacent islands.

Most cuckoos are solitary, often furtive birds that are inconspicuous even when relatively common. They do not form large flocks or even, except for the communal nesting anis (*Crotophaga*), small parties. The turacos, some of which form loose bands, are often conspicuous because of their bright colours, large size, and loud voices.

NATURAL HISTORY

Habitat utilization. As a group, cuckoos are forest birds, often inhabiting dense thickets that may make them difficult to observe. Some species, such as many of the genus *Cuculus,* inhabit rather open woodland. The guira (*Guira guira*) of South America and many members of the Old World genera *Clamator* and *Chrysococcyx* are found in open savanna (grassland), but only where trees are present. The roadrunners live in open scrub and cactus desert, often in the absence of any large vegetation.

The green and iridescent turacos (*Tauraco, Gallirex, Musophaga,* and *Corythaeola*) are primarily residents of dense, evergreen, broad-leafed forest; the grayer forms (*Crinifer*), most of which are called go-away birds (because the calls of some are "g'-way, g'-way"), are found in more open woodland, including savanna.

Food habits. Cuckoos are largely insectivorous, preying mostly on crawling insects. Orthoptera (grasshoppers, locusts, and mantids) are often taken. The greater, or North American, roadrunner (*Geococcyx californianus*) is reported to take flying grasshoppers by leaping after them from the ground. The larger cuckoos, such as the lizard cuckoos, roadrunners, and coucals, take substantial numbers of lizards, snakes, and other small vertebrates, sometimes including birds. Alone among the Cuculidae, the Madagascar couas (*Coua*) are reported to eat some fruit.

Musophagids feed almost entirely on fruit; only a few species are known to take insects and small vertebrates.

"Songs" of cuckoos

Vocalizations. Compared with birds generally, cuckoos must be considered a highly vocal group, a fact consistent with their forest habitat. A variety of songs, contact calls, and alarm notes are known for most species, some melodious, many harsh and discordant. The "song," associated with territorial assertion and courtship, is usually characterized by the repetition of loud, short notes, often on a descending scale or with a downward break in the middle. The individual syllables have variously been described as whistling, piping, cooing, tooting, laughing, grating, and clicking, depending on the species. The familiar clear, two-note call of the common cuckoo (*Cuculus canorus*) of Europe, Asia, and Africa is uttered by the male alone, the female giving a low bubbling call; apparently in most other species of cuckoos as well, the song is given by the male alone.

In addition to vocal sounds, at least one cuckoo, the North American, or greater, roadrunner, makes nonvocal sounds, clacking the mandibles to produce a rattling noise.

The calls of turacos are described as barking or laughing notes given in a series. A few species utter softer cooing notes.

Courtship behaviour. Except for features of parasitism, the behaviour of cuckoos is one of the least studied aspects of their natural history. A few species, in several subfamilies, have been observed closely, and some generalizations are possible. In territorial advertisement and in courtship, the singing male (in both parasitic and nonparasitic cuckoos) may posture, spreading the tail feathers, which are white tipped in most species, and drooping the wings. Courtship feeding of the female by the male has been observed in several species.

Courtship behaviour has been described for a few musophagids. In forest species, the male walks or leaps among branches with his tail fanned and with his wings half spread, displaying the crimson wing patches. Courtship feeding has also been observed, the male regurgitating fruit pulp or presenting whole fruit to the female.

Nesting. Two basic types of nesting behaviour are found among members of the order: parasitic and nonparasitic. About three-fifths of the Cuculidae and all of the Musophagidae mate, nest, and rear their young in the manner seen in most birds. About 50 species of cuckoos, including all of the subfamily Cuculinae (sometimes called "typical" cuckoos) and three species of the Neomorphinae, exhibit brood parasitism (*i.e.,* they lay their eggs in the nests of other species, which then rear the young cuckoos). Brood parasitism in the common cuckoo was recognized by Aristotle in the 4th century BC, and that of the koel (*Eudynamys scolopacea*) appears in the ancient Vedic literature of India, dating from about 2300 to 1800 BC, but the scientific study of this form of parasitism remained largely neglected until the late 19th century, when intensive study by many observers revealed a number of remarkable behavioral and physiological adaptations in the cuckoo.

Egg laying by the common cuckoo

In western Europe the common cuckoo is migratory, arriving at its breeding area in spring. The female establishes a territory and defends it against intrusion by other female cuckoos. The male is less restricted to an individual territory and more tolerant of other males; often more than one male will be in attendance on one female. Moving about within her territory, the female learns of potential host nests largely by watching their construction. Males of some species of cuckoos may accompany the females in searching for potential host nests. By the time she is ready to lay her eggs the female cuckoo has under surveillance several possible host nests and knows when the first eggs are laid by their owners. Most of the species parasitized by the common cuckoo are smaller than the cuckoo and lay their eggs at the rate of one each day until the clutch is complete, only then beginning incubation. Visiting when the hosts are away from the nest the cuckoo is able to place her egg in the nest after the host has started the clutch but before incubation has begun. It is well established that the common cuckoo lays eggs at 48-hour intervals, and apparently some glossy cuckoos lay one every 24 hours. The cuckoo places only one egg in each host nest and so must find a new nest for each of her 15 to 20 eggs. When parasitizing a species that builds a domed or globular nest, the female cuckoo clings to the outside of the nest and lays her egg through the opening without entering the nest. The repeated finding of cuckoo eggs in nests too small to admit a cuckoo or to bear her weight has led some authorities to maintain that the cuckoo lays the egg on the ground and carries it to the nest in her beak, but extensive observations have failed to substantiate this supposition. Some of the glossy cuckoos, however, parasitize nests so small and frail that mandibular egg placement cannot be ruled out.

In western Europe common cuckoos lay their eggs in the nests of many species, mostly songbirds (Passeriformes); more than 140 hosts species have been reported, including some that could hardly rear a young cuckoo successfully. Careful study of this and several other cuckoos indicates, however, that despite the wide range of hosts recorded, a few host species bear most of the load of raising the young of each cuckoo species. In some instances, such as two of the little Australasian glossy cuckoos (*Chrysococcyx*), the parasite relies mainly on a single host species.

Egg mimicry

Many birds react to a foreign object in the nest by deserting the nest, building another nest on top of the first, or removing the offending object. It is therefore of great advantage to a brood parasite to lay eggs that resemble those of its host, or, conversely, to parasitize species whose eggs are the same colour as its own. The utilization of several host species by a single population of cuckoos presents a difficult adaptive problem, for the host species may have many different egg colours. Early naturalists noted that there was often a marked resemblance between the egg of a cuckoo and those of the host, and a German ornithologist, Eduard Baldamus, in 1892 showed that the frequency and degree of similarity were too great to be coincidental. Subsequent studies by a number of workers, especially by an English naturalist, Edgar P. Chance, have revealed much of the basis for the resemblance, which is now called egg mimicry.

Each female common cuckoo lays eggs of a particular shape and coloration throughout her life, but the eggs of different females vary widely. Most authorities now agree that a given female normally restricts her parasitism to the single host species by which she herself was reared, and that the common cuckoo population is composed of numerous clans, called gentes, each of which parasitizes only one species of bird, for which the females of that gente have evolved egg mimicry. Only the gente of the female is important in egg coloration; the foster parentage of the male (*i.e.,* his gente) has no effect on his choice of females, and his genetic makeup apparently does not influence the coloration of eggs laid by his female offspring. Gentes are not given taxonomic designations but are sometimes known by their host species ("redstart cuckoo," "dunnock cuckoo," etc.). Evidence for this theory is largely circumstantial because of the difficulty in establishing the identity of individual cuckoos, but no plausible alternative theory has been proposed. The presence of cuckoo eggs in a wide variety of host nests in some small areas has been taken as evidence refuting the single-host hypothesis, but most of the nests with nonmatching eggs are found in areas in which the ecology has been seriously altered by human activity, with the result that female cuckoos have suddenly found themselves without adequate numbers of the proper host species and have been forced to utilize hosts for which their own egg colours are inappropriate. In the forested portions of northern Europe, where the habitat is relatively undisturbed and consists of homogeneous microecological areas (*e.g.,* reedbeds, birch forests), the common cuckoo parasitizes mainly a small thrush, the redstart (*Phoenicurus phoenicurus*), and both lay unspotted bluish eggs; in the reedbeds around Hungarian lakes the usual host is the great reed warbler (*Acrocephalus arundinaceus*), and the cuckoo eggs are strongly blotched with gray and black, like those of the warbler.

Studies of the reaction of the host to the presence of a cuckoo egg have indicated that egg mimicry is of considerable importance to the success of the parasite. Only about 5 percent of well-matched eggs are rejected by their hosts, as compared with up to 72 percent of mismatched eggs. Few cuckoos have been studied intensively in terms of egg mimicry, but the phenomenon is known to occur in at least some species. The great spotted cuckoo has an egg pattern mimicking that of the magpie (*Pica pica*), its usual host in southern Europe. In Africa, where it is apparently a recent colonist, this cuckoo exhibits what has been termed "evolutionary escape from specialization"; its new hosts, certain starlings, nest in holes, in which the semidarkness renders the dissimilarity of eggs less visible and hence unimportant. Among the 12 species of emerald cuckoos, the degree of host specificity and egg resemblance is variable, some species being highly host adapted, others less so. The genus as a whole shows no clear continuity of evolution in egg adaptation; individual species evidently have evolved their egg mimicry separately.

A notable feature of the egg of the common cuckoo is its small size; its weight is only about 2.5 percent of that of the adult cuckoo. The eggs of other parasitic cuckoos weigh more than 6 percent of the adult weight, and those of birds that rear their own young vary from 8 to more than 20 percent of adult weight. The adaptive value of small eggs is not well understood, but it is assumed to be related to the reliance of the common cuckoo on small hosts. Recent experiments have shown, however, that incubating birds show no adverse reaction to eggs significantly larger than their own.

Nest-mate eviction

Once the egg has been accepted and incubated by the hosts, the newly hatched cuckoo faces the problem of securing enough to eat from its small foster parents. The young cuckoo has evolved an astonishing form of behaviour, that of nest-mate eviction, that ensures that it will not have to compete with members of the foster brood for food. Within a few hours of hatching, the blind, naked, young cuckoo develops a strong urge to evict any objects, such as eggs or other nestlings, from the nest. It does this by working itself under the offending object and, aided by the presence of a depression between the shoulderblades, heaving the object over the rim of the nest. Within about 24 hours of hatching, the young cuckoo has the nest and the attentions of its foster parents to itself. The eviction habit is well developed in some parasitic cuckoos and seems to operate effectively in favour of the evictor but is surprisingly absent in many others, such as the crested cuckoos (*Clamator*).

Another habit that, like nest-mate eviction, tends to reduce the competition faced by the young cuckoo is the removal of one or more of the host's eggs by the female cuckoo, sometime before or after or at the time of the laying of her own egg.

The young cuckoo, while still in the nest, grows rapidly in response to the efforts of its foster parents, who regard it as their own. By the time it is ready to leave the nest, the young bird dwarfs its smaller hosts. Normally it does not receive attention from its real parents, but adult common cuckoos have occasionally been observed feeding young cuckoos; in the genus *Chrysococcyx* such behaviour appears to be more frequent than in other genera.

Social nesting in the guira and anis

Most of the nonparasitic cuckoos form stable pair bonds and defend territories, within which they build their nests and rear their own young. The guira and the anis are exceptional in that they live in flocks of five to 20 individuals, each flock defending a territory within which its members feed and nest. Several birds of the flock may cooperate in building a nest, in which two or more females may lay eggs and share incubation. Many members of the flock participate in feeding the young.

The nests of nonparasitic cuckoos are loose platforms of twigs placed in low vegetation or, rarely, on the ground. The coucals are unusual in that they build sizable domed nests of grass and twigs, on or near the ground, with side entrances.

The musophagids also build large twig nests, placed in trees often at considerable height. They lay unmarked whitish eggs, from which the young are believed to hatch completely covered with down; certainly they are downy from an early age.

Migration. Like most insectivorous birds in temperate climates, temperate-zone cuckoos migrate toward or across the equator for the winter. The migrations of some species are remarkable in that the young of the year travel completely independently of their parents and may cross up to 3,200 kilometres (2,000 miles) of open ocean unguided. Adults and young of some of the small Australasian glossy cuckoos separately migrate long distances to relatively small oceanic islands. It is obvious that the means of orientation and navigation must be unusually precise, but what they are remains as yet unknown.

FORM AND FUNCTION

In general body plan, most cuckoos resemble the "perching birds" (order Passeriformes). They are slender bodied, with medium to long wings and medium to extremely long tails, the latter sometimes constituting more than half the total length of the bird. The tail is usually graduated (the outer feathers shorter), and the individual feathers often are white-tipped.

The foot structure of cuckoos and musophagids is distinctive and at close range distinguishes these birds from any with which they might be confused. The foot of cuckoos is zygodactylous, or yoke-toed (*i.e.,* the outer toe is directed backward); that of musophagids is semi-zygodactylous, the outer toe being movable into either a forward or a backward position.

The cuckoo bill, although usually slender, is occasionally quite stout, as in the ground cuckoos and coucals, in which it makes a formidable weapon for subduing prey. In the anis the bill is heavy, with a strong ridge along the culmen (top), and, in the channel-billed cuckoo (*Scythrops novaehollandiae*) of Australasia, it is extremely large, almost like that of toucans. The bill is strong and rather short in the Musophagidae, usually with serrated edges. In the genus *Musophaga* it extends onto the forehead as a broad shield.

Most musophagids have prominent crests, a feature rare in cuckoos and well developed only in the genus *Clamator.* A number of cuckoos and turacos have bare skin, usually brightly coloured, around the eye or between the eye and the bill.

Turaco feather pigments

Perhaps the most distinctive feature of the turacos is the presence of two unique feather pigments. Turacin, a rich crimson pigment found in the primary flight feathers of *Tauraco* and *Musophaga,* is a copper salt of the pigment uroporphyrin III; and turacoverdin, the green pigment in turacos, is an oxidized derivative of turacin.

PALEONTOLOGY AND CLASSIFICATION

The order Cuculiformes is not well represented in the fossil record, but both cuckoos and musophagids apparently existed by the Early Tertiary Period, some 60,000,000 years ago. The dearth of fossil evidence is not surprising, in view of the group's preference for forest habitats; woodland animals, particularly small ones and particularly birds, are less frequently fossilized than those in open areas or aquatic habitats. Two species of musophagids have been described from the upper Eocene or lower Oligocene of France (about 40,000,000 years ago) and one from the middle Miocene (about 15,000,000 years ago). Three species of fossil cuckoos are known from North America, the earliest from the middle Eocene (about 45,000,000 years ago), and a fossil coua has been found in deposits from the Quaternary Period (the last 2,500,000 years) of Madagascar.

Although the cuckoos and musophagids are frequently united in a single order, as in the present article, most authorities acknowledge the distinctness of the two groups by placing them in separate suborders, Musophagi and Cuculi. The parrots (Psittaciformes) are often considered on osteological evidence to be the closest relatives of the cuculiforms, but even this relationship, if it exists at all, is not a close one. (He.F.)

Strigiformes (owls)

Owls constitute a homogeneous order (Strigiformes) of primarily nocturnal birds of prey, nearly worldwide in distribution. Their secretive habits, quiet flight, and weird calls have made them the objects of superstition and even fear in many parts of the world. The same characteristics have made the scientific study of them more difficult, with the result that the ecology, behaviour, and taxonomy of many species remain poorly understood.

GENERAL FEATURES

The size range of owls is about the same as that of their diurnal counterparts, the hawks (exclusive of the eagles and vultures), with lengths of around 13–70 centimetres and wingspreads around 30–200 centimetres. There are more owl species at the lower end of the size range. Owls apparently feed exclusively on animals. The most common prey is rodents; the smaller species, however, eat insects. All owls have the same general appearance: large head; hooked beak; large, forward-looking eyes; short tail; rather long, rounded wings; and large feet with sharp talons. Diversity occurs in size, in the presence or absence of decorative "ear" tufts, and in the extent of the facial ruff.

Owls are found on all continents except Antarctica and on most oceanic islands. Some, such as the barn owl (*Tyto alba*) and the short-eared owl (*Asio flammeus*), are among the most widely distributed of bird species; others, such as the Palau owl (*Pyrroglaux podargina*) and the Seychelles owl (*Otus insularis*), are endemic island species with small populations. Owls often attain higher population densities than hawks and have survived better in areas of human activity. Their nocturnal habits and inconspicuous daytime behaviour provide them some protection from shooting. The greatest densities are attained by small, territorial, insectivorous species, with pairs spaced about 200 metres apart in suitable woodland.

NATURAL HISTORY

Ecology. Owls utilize virtually all habitats, from grassland and tundra to dense woodland and rain forest. The distribution and density of most species apparently are limited by the availability of suitable nesting sites, rather than by the number of potential prey animals. In general the type of prey taken is governed by the size of the owl and by the relative abundance of potential prey species.

Hunting habits

Owls that hunt over grassland, such as the barn owl and short-eared owl, hunt by sustained flight, dropping into the grass to catch rodents. Many woodland owls secure prey by dropping from perches at the edges of forest openings. The Southeast Asian hawk-owl (*Ninox scutulata*) sallies from a perch to take flying insects. The whiskered owl (*Otus trichopsis*) takes flying insects in foliage. Owls of two genera (*Ketupa* and *Scotopelia*) are adapted for taking fish. Specialized forms of feeding behaviour are sometimes observed. The elf owl (*Micrathene whitneyi*) has been seen hovering before blossoms, scaring insects into flight with its wings and then catching them with its beak. A bay owl (*Phodilus badius*) has been observed stationed within a bat cave, catching bats as they issued forth at dusk. Although each species normally feeds on the type of prey for which it is adapted, a variety of owls may depend on a single prey animal when it becomes exceptionally abundant. Prey is generally swallowed whole, and indigestible material, such as feathers, fur, and bones, are regurgitated in the form of a compact pellet.

Vocalizations

Behaviour. Sound communication is important to owls, especially in mating and territorial defense. With their camouflage, immobile daytime stance, and silent flight, owls may be as difficult for each other to see as for natural enemies and human observers. Their characteristic sounds include snaps of the bill, claps of the wings in flight, and a variety of vocalizations, with pitch, timbre, and rhythm unique to each species and with pitch, at least, differing between sexes (the female higher). Although less melodious than the calls of some birds, the vocalizations of many owls are "songs" in the biological sense and may even be musical to the human ear. The song varies from deep hoots in some large species to chirps, whistles, or warblings in many small owls. When the nest is threatened, nestling burrowing owls (*Speotyto cunicularia*) emit a buzzing call that resembles the warning buzz of the rattlesnake, a frequent inhabitant of rodent burrows.

In the North American screech owl, a duet that seems to reinforce the pair bond starts with a special song by the male, eventually answered in kind by the female, often from a distance. After ten to 15 minutes of antiphonal (answering) singing, during which the two approach each other, the pair switches to a duet with a second song pattern, during which they meet. In the early spring this may be followed by precopulatory calls and posturing, then mounting. Other calls of the screech owl include: a note uttered by the female to stimulate the young to reveal their location after they have left the nest; a food soliciting call by the young; after the breeding season, barking calls, accompanied by bill-snapping, which indicate that the young are being ejected from the territory, or the adjustment of territorial boundaries.

Drawing by R. Keane (bay owl) and D. Malick (others)

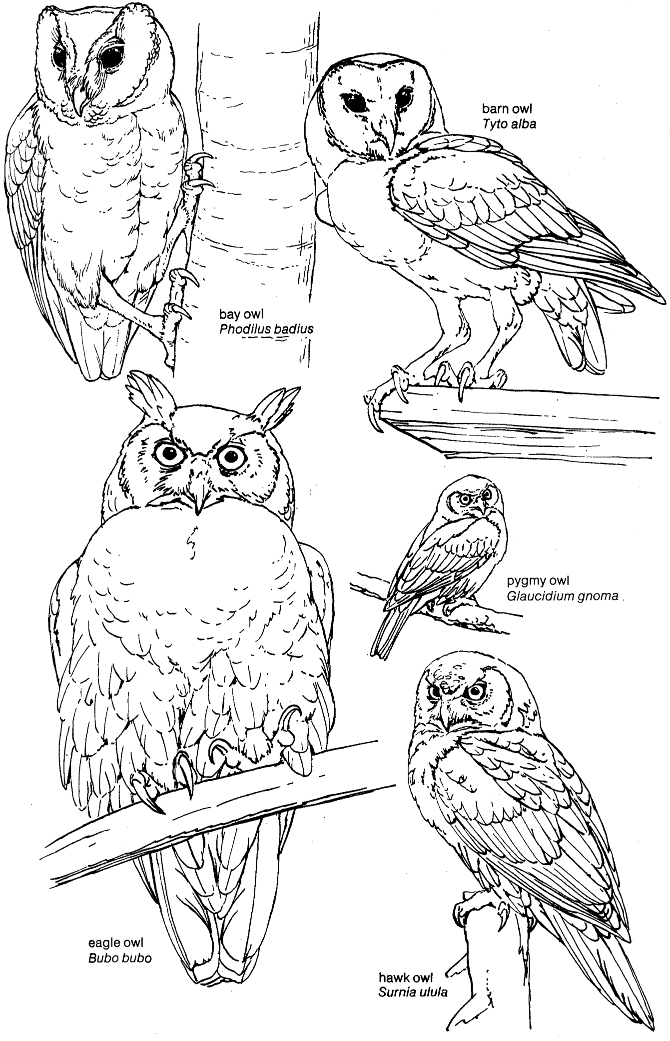


Figure 36: Variations in body plan among owls.

In many smaller species that do not normally sing duets, the male may sing all night from a single perch. The nocturnal routine of most owls involves peaks of activity at dusk and dawn. The owl leaves its secluded roost about dusk and moves to a perch overlooking the hunting area. There is a brief period of song, followed by about half an hour of foraging, then a longer period of song. Most of the darker hours of the night are spent inactively, with a period of alternating singing and hunting just before dawn.

Posturing by singing owls indicates that they communicate by sight as well as by sound. The male horned owl bows deeply with each song and raises the tail over the back. The wood owls (*Strix* species) engage in bowing, bobbing, and dancing, especially when courting. A defense display given by most large owls when threatened or when defending the nest involves increasing apparent body size by half spreading the wings and rotating them forward. The body feathers are raised and the fearsome appearance enhanced by snapping the bill and rocking the body from side to side. When seeking to avoid attention at its daytime roost (especially when being attacked by small birds), the owl compresses the plumage, elevates the ear tufts, and half closes the eyes. Combined with the barklike colour pattern of the plumage, this behaviour provides the appearance of a broken branch.

Nesting sites

Reproduction and development. Most owls nest in natural cavities in trees or cliffs or in woodpecker holes. The barn owls and the Eurasian little owls (*Athene* species) frequently use cavities in buildings. Some of the larger owls utilize old hawk or crow nests. Grassland and tundra owls nest on the ground, sometimes on an elevated hummock, and the burrowing owl digs a nest chamber in a rodent burrow. Frequently the nest cavity provides a daytime roost for one or both of the pair during the nonbreeding months. Most owls add no nesting material to the site, but the fur and feathers of accumulated prey remains and regurgitated pellets may provide some cushion for the eggs. When an open nest is used, leaves, grass, or other soft material may be added as a lining. The great gray owl (*Strix nebulosa*) occasionally constructs its own platform nest in a tree. In desert areas the smaller owls rely primarily on holes made by woodpeckers in large cacti. Intense competition has been observed among nesting birds, including owls, for occupancy of a limited number of nest sites. The invasion of the saguaro desert by the European starling (*Sturnus vulgaris*) has had a serious effect on small owls and other saguaro birds. The aggressive and abundant starlings occupy cavities before the other species have returned from wintering grounds and successfully defend the holes against native species.

Egg laying is timed so that the young become independent of their parents at a time when populations of prey species are greatest. At northern latitudes, many owls nest a month or two earlier in the spring than hawks of the same regions. An incubating owl is frequently covered by several inches of snow. Barn owls have been found nesting in every month of the year, even at the northern edge of their range, but the peak of nesting is in the spring. Owls lay more eggs than most diurnal raptors, with clutches of up to 12 (in the snowy owl, *Nyctea scandiaca*). In years of lemming abundance, snowy owls attain higher nesting densities, nest earlier, lay more eggs, and have higher fledging rates than when lemmings are scarce. The eggs of owls are more nearly spherical than those of any other bird group, the long diameter averaging only about 1.2 times the short diameter. They are laid at one- or, more typically, two-day intervals, and incubation begins when the first is laid. The result is that the oldest and youngest nestlings of a large brood may be hatched two or three weeks apart. If the prey populations are inadequate for the adult owls to support the entire large brood, the younger nestlings starve, while the more aggressive older ones are able to maintain normal growth rates and are strong at fledging.

Plumage succession

While in the nest, young owls grow two successive coats of white natal down. In the smaller species, the down is replaced by the immature juvenal plumage of softer, lacier texture than that of the adult. At this age, about two-thirds of the way to fledging, the young owls may leave the nest and spend the day several metres from it. In the screech owl, the juvenal plumage has fine barring, unlike the streaking of the adult. The first flight feathers, which appear about the same time, are like those of the adult but are more pointed. In the fall, only a few weeks after its acquisition, this plumage is replaced in a complete molt that introduces a plumage identical to that of the adult. In the larger owls that have been extensively studied (several species of *Bubo*), the juvenile plumage resembles the adult except that in *B. nipalensis* and *B. sumatranus* it is mostly white, with a different pattern of black marking from the adult.

Locomotion. Most owls use their feet only for perching and grasping prey. The burrowing owls, being terrestrial, can run rapidly over the ground and rarely perch in vegetation. The typical perch of other owls is a horizontal branch in a tree. The bay owl habitually perches on the vertical trunk of a tree, with the legs reaching sideways, one foot above the other. The body is held in a vertical position as though the owl were perched on a horizontal limb.

The flight of owls is a steady flapping on a straight path, ending in a short upward glide to the perch. Hunting, usually done from a perch, rarely involves extensive flight, requiring only a short burst of speed to surprise the prey on the ground. Forest owls rarely fly above the canopy of the foliage, but the Southeast Asian hawk-owl has been observed high in the air, flapping—or, on one occasion, soaring—in circles through swarms of bats, apparently without catching any. A few grassland and tundra owls hunt in sustained flight and have proportionately more wing area than their forest-dwelling relatives. The short-eared owl flaps slowly, the large area of the wings causing the light body to bob up and down, and sails for a brief period of time with the wings held in a high V over the back.

FORM AND FUNCTION

All owls share the same general body plan. The wings are long and rounded, the tail short. The legs and toes are of medium length and exceptionally strong for the size of the bird. Each toe is provided with a needle-sharp, curved talon. The outer toe is reversed in perching and is normally directed outward or backward in taking prey, providing the maximum possible toe spread.

The head is broad, to accommodate the exceptionally large eyes. The owl eye is elongated forward and encased in a tube made up of joined bony elements. The eye is virtually immovable, rigidly encased in the bony socket. Remarkable flexibility of the neck compensates for the immobility of the eyes; an owl can turn its head more than 180° in either direction and can thus look directly backward. The vision is binocular, and depth perception is often enhanced by moving the head away from the central plane. Owls that are mostly active at night have only rods in the retina, resulting in an absence of colour vision but a great increase in visual acuity and light sensitivity. Contrary to popular opinion, owls are not blind in strong light. The pupils, which operate independently, can be greatly reduced, protecting the sensitive retina and providing better daytime vision than that found in man.

Hearing

The ear is large and is surrounded by a ruff of papery feathers that serve to concentrate the sound. The feathers covering the ear opening are lacy and permeable to sound. A movable flap (operculum) on the anterior margin of the opening may function as a baffle to focus sounds. Some owls can locate and capture prey in total darkness, relying on their ability to localize the rustle of a mouse in leaves and to fly to that spot. In many owls the relative position of the ear opening is asymmetrical, being above a so-called blind cavity on one side of the head and below it on the other. The asymmetry is thought to be related to the sensitivity of each ear to sounds of various frequencies and provides the owl with the ability to localize sound sources in two planes simultaneously.

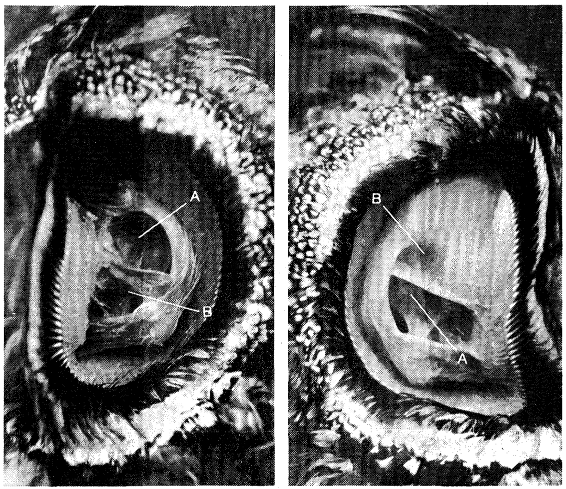
Many owls have erectile tufts of feathers ("ears" or "horns") above the eyes. In some cases the tufts give the owls a feline appearance and might frighten potential predators, but it is more likely that the erect "ears" serve to break the rounded outline of the head, adding to the concealment gained from colour and pattern.

The plumage of owls is soft, dense, and loose. A thick layer of down provides northern owls with insulation against cold. The upper surfaces of the flight feathers of most species are provided with a nap that makes the flight perfectly noiseless, allowing the owl to hear prey without interference of flight sound.

Owls vary in coloration from white through many shades of tan, gray, brown or rufous to deep brown. A few are solidly coloured, but most are cryptically patterned, with streaks, bars, or spots, often resulting in the bird's being almost invisible against bark. The concealing bark pattern is well exemplified in the small owls of the genus *Otus*. The soft brown, rufous, buff, or gray ground colour of each breast feather is adorned by a blackish bar, shaft streak, or combination of both, sometimes outlined in white or rufous. In some widespread species, such as the Eurasian scops owl (*O. scops*) and the screech owl, geographic variation is so great that the more divergent races are more different than some full species. In the far north there is only a faint pattern on a whitish background; in humid, temperate forests a bold pattern on a sooty background; in desert areas a medium to fine pattern on pale gray; in arid tropics a fine pattern on rufous; and in humid tropics a coarse pattern on fulvous. Size variation is also present, with northern birds weighing about twice as much as those of southern races. The horned owl (*Bubo virginianus*) exhibits similar variation. Screech, scops, and whiskered owls may be either gray or rufous; the ground colour apparently is determined by a single gene. Such dimorphism in ground colour is found in only certain populations of each species: the southern in the whiskered owl and the eastern in the North American screech owl and the Eurasian scops owl. In each case, interbreeding between the uniformly coloured monomorphic and sexually distinct dimorphic populations is limited. J.T. Marshall, Jr., regards the dimorphic strains as "incipient species." The mechanism by which natural selection maintains colour dimorphism, *i.e.*, its selective advantage, is poorly understood. Moreover, little is known about the type and degree of predation on small owls. The red phase of the screech owl may have survival value in the predominantly deciduous forest of eastern North America, blending with the foliage of summer and autumn, in which reds and browns are strongly represented.

Colour dimorphism

Heinz Meng



Asymmetry of the ear canal in the long-eared owl (Asio otus). In the left ear (left), the opening of the ear canal (A) is above the transverse fold and the blind cavity (B), below. In the right ear, the situation is reversed.

PALEONTOLOGY AND CLASSIFICATION

Fossil history. The earliest recognizable fossil strigiform is *Protostrix* from the Eocene of Wyoming. A number of genera from the later Tertiary have been assigned to Recent strigiform families. Some early owls reached far greater size than their modern descendents. A giant barn owl, about twice the size of the modern *Tyto alba,* inhabited Puerto Rico in the Pleistocene. Another large owl, *Ornimegalonyx oteroi* from the Pleistocene of Cuba, apparently was flightless. Both owls must have exceeded the modern eagle owls in size.

Distinguishing taxonomic features. The owls form a homogeneous group readily distinguished from all other orders by their general body plan, soft plumage, and osteological peculiarities. To determine relationships within the Strigiformes, taxonomists utilize the osteology of the skull and the sternum, relative specializations of the eye and ear, and the development of the facial disk. Within some genera the taxonomy is very complex, relying on general proportions, behaviour, voice, and ectoparasites (Mallophaga or feather lice).

Annotated classification.

ORDER STRIGIFORMES

Nocturnal predatory birds with hooked beaks; nostril marginal to fleshy cere; broad skull, facial disk, with forwardly directed eyes; reversible outer toe; talons for securing prey. One fossil and 3 extant families.

†Family Protostrigidae.

Fossil. Known only from lower Eocene of Wyoming.

Family Tytonidae (barn owls and grass owls).

Lower Miocene to present. Medium-sized owls (30–54 cm long) with heart-shaped facial disk completely encircling face; bill relatively long and slender; legs rather long, middle claw with comb. One pair of sternal notches, clavicles separate and not fused to sternum. One genus (*Tyto*), about 9 species. Worldwide, except in polar and subpolar regions and on some oceanic islands.

Family Phodilidae (bay owls).

Miocene and Recent. Like a small *Tyto,* but clavicles fused to sternum, which has 2 pairs of notches. Facial disk differs from *Tyto* in forming 2 movable lappets that can be focussed at sound, like mammalian ears. Disk interrupted at forehead. Wings short, broad. Legs and toes shorter and thicker than in *Tyto.* Voice, syrinx, behaviour, and ecology differ from both Tytonidae and Strigidae. One Recent genus (*Phodilus*), with 1 species widespread in Southeast Asia, and another (known only from a single specimen) in central Africa.

Family Strigidae (typical owls).

Upper Eocene to Recent. Clavicles fused to sternum, which has 2 pairs of notches. Facial disk consists of a semicircle on each side. Middle claw without comb. About 22 Recent genera, 125 species. Length about 13–70 cm. Worldwide, except Antarctica and some oceanic islands.

Critical appraisal. Most recent taxonomists have divided the Recent Strigiformes into two families, Tytonidae and Strigidae. The bay owl (*Phodilus*) has been placed in the Tytonidae by some authorities and in the Strigidae by others. F.E. Beddard, in 1898, suggested that the distinctive characteristics of the bay owl justified separation in a monotypic family, Phodilidae, a view recently endorsed by J.T. Marshall, Jr., after observations in the field.

Within the Strigidae, most current taxonomic problems are concerned with the placement of certain genera within the family and with the specific status of some populations of complex genera, such as *Otus.* Some of these problems await the development of new information on species-isolating mechanisms, especially vocal behaviour. (J.T.M.)

Caprimulgiformes (nightjars, frogmouths, potoos, oilbird)

The order Caprimulgiformes includes about 100 species of soft-plumaged birds, the major groups of which are called nightjars, nighthawks, potoos, frogmouths, and owlet frogmouths. The order also includes the aberrant oilbird of South America. Most are twilight or night-flying birds. Many produce sounds that are startling, strange, or weirdly beautiful and are surrounded by an aura of mystery richly endowed to elicit interest and sometimes fear from man. The name of the type genus *Caprimulgus,* "goatsucker," derives from an ancient belief that the birds seen flitting about the goats at dusk, actually preying on the insects disturbed or attracted by the goats, were taking milk from the goats' udders, a misconception no doubt fortified by the birds' uncommonly large mouths. There is now a tendency to replace the name goatsucker with the more appropriate term nightjar, derived from the birds' voices.

The caprimulgiform birds are sparrow- to raven-sized (14–55 centimetres) birds with enormous gapes, cryptically coloured and patterned plumage, short legs and, for the most part, long wings.

Comparison with owls

All caprimulgiform birds are rather similar in general appearance, but each family has certain peculiar characteristics both in form and in habits. Their closest relatives are the owls (Strigiformes), which they resemble in many ways, but there are numerous differences between them. Many of these are internal, but among those externally apparent are the bill and feet, which are not raptorial; the flatter head with eyes placed laterally rather than in a frontal facial disk; the relatively shorter tarsi and longer tails. Closer inspection reveals differences in the number of primary feathers in the wing and usually of secondaries and tail feathers as well. The nightjars also share some features with the swifts (Apodiformes), but these seem to be more superficial and coincidental than indicative of close phylogenetic relationship.

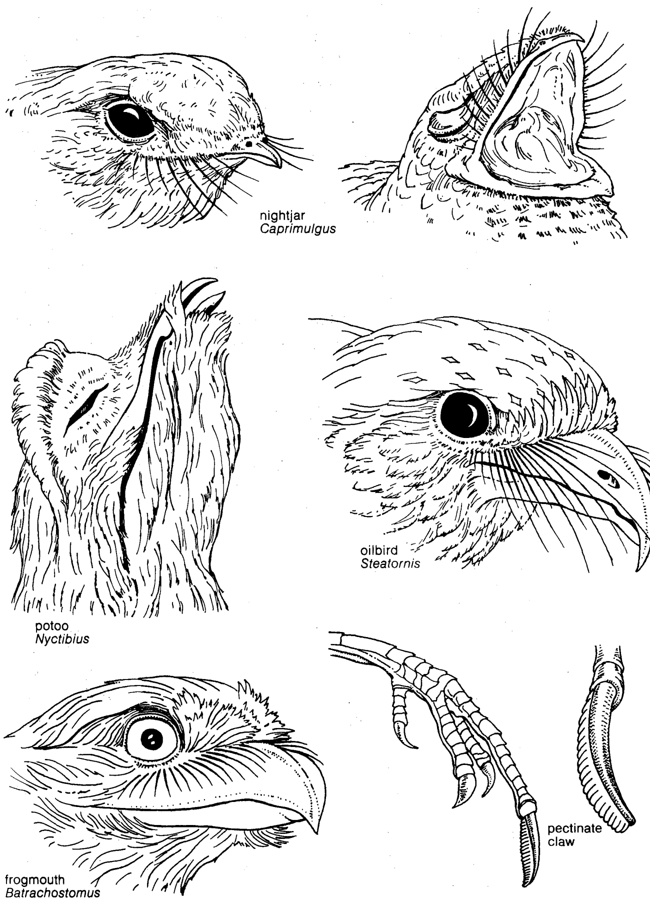


Figure 37: Beak and claw structure of representative Caprimulgiformes.

Drawing by E.J. Pfiffner

Although the true nightjars (Caprimulgidae) are amply distributed throughout the world, the other families are more restricted. The order is absent from New Zealand and some oceanic islands.

NATURAL HISTORY

The caprimulgiform birds are primarily crepuscular, their activity being largely limited to the periods of dawn and dusk, although they are also nocturnal when there is sufficient illumination, especially by moonlight. Some species may become active on dark, cloudy days, and a few are somewhat diurnal. Oilbirds possess a system of echolocation that permits them to fly freely in total darkness, an adaptation related to their roosting and nesting in caves, and are thus implicitly well equipped for nocturnal life. They leave their caves at dusk and return at dawn and are highly gregarious in their foraging behaviour as well as in roosting and nesting. The other members of the order are more or less solitary. Some nightjars that are migratory behave gregariously during migration and to some extent while in their "wintering" regions.

Although aerial feeders, most of the true nightjars roost on the ground, rocks, or fallen trunks, but some prefer horizontal branches of trees, in which case they usually perch lengthwise along the branch. Some ground roosters, however, will seek higher perches as singing or foraging posts; these are often slender branches or vines and the birds sit on them crosswise. Some species may even roost so perched. Unlike the nightjars, the frogmouths, potoos, and owlet frogmouths are arboreal. The last normally sit crosswise on a branch and fairly upright, both when active and at rest, resembling small, long-tailed owls. Potoos and frogmouths frequently sit crosswise and upright when actively foraging and also appear very owllike, although at rest they may perch quite differently. The potoos are noted for a peculiar stance, perching usually at the top of a stump, broken branch, or at a knob on an upwardly inclined limb. When alarmed in daylight they slowly flatten their plumage and stretch their bills upward in a stiff posture with the eyes nearly closed and the bill slightly open. Frogmouths adopt a similar broken-branch alarm posture.

During the day, the oilbird perches horizontally on ledges inside caves, usually on its nest. Ledges are limited and mostly occupied by nests. On its forays outside it has been observed to perch on the bare parts of tree branches. Its mien is rather hawklike, but its crouching stance, unique in some ways, is more like that of a nightjar.

Habitat. The caprimulgiform birds exploit ample food sources that are almost uniquely theirs. There are no other primarily frugivorous nocturnal birds to compete with the oilbird, its distribution being limited by the availability of suitable caves in regions providing the proper fruit for food.

Competition between species

The nightjars, by far the largest and most successful group, have no effective avian competitors for the great numbers of night-flying insects except among themselves; and the different nightjars have come to occupy almost all habitats, from semidesert to very humid regions and from sea level to 4,000 metres in altitude, with further advantage by preferring the more open areas: savannas, grasslands, and forest clearings. When woodland dwellers, they live in the more open forests or brushwood growths, and many use the woodlands only for diurnal concealment or nesting, doing most of their feeding at the edge or in clearings.

Competition between the nightjars and potoos is reduced because the latter hunt mostly from higher perches in open country or exploit the space above the great forests, a vast niche occupied by practically no other nocturnal bird.

Frogmouths and owlet frogmouths inhabit forested savannas, forest edge, second growth, and the more open forests. Thus their chief competitors are the insectivorous owls. As with the nightjars, differences in food preferences and methods of hunting minimize this competition. The hole-nesting owlet frogmouths, however, have to contend not only with owls but also many diurnal species for suitable nesting sites, and this may explain why they are apparently the least successful family in the order, as suggested by their limited distribution and paucity of numbers.

Food habits. Except for the oilbird, caprimulgiform birds are insectivorous, with some of the larger species occasionally carnivorous. Many of the nightjars and nighthawks feed in continuous hawking flight, which may be rather erratic as they pursue their prey. Others of the order make short sallies from arboreal or terrestrial perches in the manner of flycatchers, frequently taking advantage of the increased visibility afforded by roads. Many nightjars take crawling insects by swooping from arboreal perches or by running along the ground, being far more agile on their feet than is generally believed. Almost all insects are taken, but beetles predominate in the food of many species. Also important are moths, the winged forms of termites and ants, and mosquitos; even small birds are taken. There seems to be little difference between the general kinds of food items taken by the New and Old World nightjars (subfamily Caprimulginae), but the Old World species rely more on foraging in continuous flight than do most in the New World, where aerial foraging is largely limited to the nighthawks (Chordeilinae).

Potoos sally from exposed arboreal perches, sometimes fairly close to the ground, to feed on essentially the same types of insects as the nightjars. The prey of frogmouths is mostly terrestrial. In short sallies from elevated perches, frogmouths capture large, crawling arthropods on the ground or on branches. Snails, frogs, mice, small birds, and occasionally fruit also are taken. Owlet frogmouths feed mostly on terrestrial prey, much in the manner of the frogmouths, but also sally from a perch to take flying insects.

The frugivorous oilbirds

Aberrant in so many ways, oilbirds are the only nocturnal, exclusively frugivorous birds. The fruits eaten are primarily of various trees of the palm, laurel, and bursera families and have large hard seeds. The firm, fleshy pericarp of the fruit is exceedingly rich in protein and oily fats. The fruits are swallowed whole and the seeds later regurgitated. D.W. Snow has found that oilbirds apparently possess an efficient olfactory sense, suggesting that they employ scent in locating certain of the fruit-bearing trees. The fruits of palms, however, are not aromatic and are no doubt located by sight. The actual collection of the fruit seems to depend on sight. In feeding, the birds reach out with their bills and pluck the individual fruits while hovering. When feeding on large compact bunches of fruit they sometimes cling briefly with their feet.

Males of most, if not all, nightjars have patches of white feathers that serve as signals in mating. These patches are concealed when the bird is at rest but flashed in courtship displays. Males of several species in Africa and South America develop elongated wing or tail feathers that function in courtship and are lost following the breeding season.

Drawing by E.J. Pfiffner

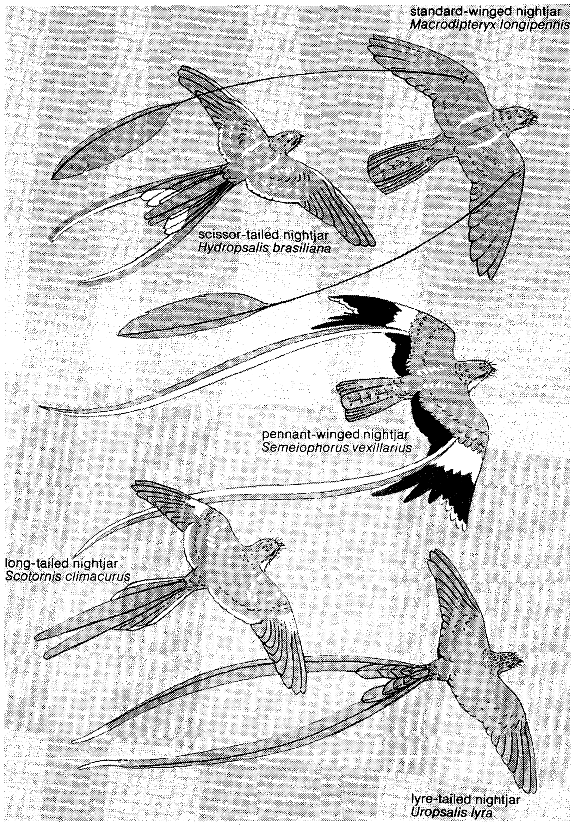


Figure 38: Specialized feathers in male Caprimulgiformes.

Vocalizations. The primary signals for mating and territorial purposes in most caprimulgiforms, however, are vocal and mechanical sounds. The voice is usually well developed, often having such distinctive patterns that the birds are named onomatopoeically (whippoorwill, chuck-will's-widow, etc.). The almost human lament of the common potoo (*Nyctibius griseus*) in South America has in some places earned this species the name "poor-me-one," and both it and the grotesque bawl of the great potoo (*N. grandis*) have been the source of many superstitious beliefs and legends. Frogmouths sing with various nasal booming, hooting, or croaking sounds, and the little known owlet frogmouths are reported to give owl-like whistles, churring calls, and a loud hissing note. Among the true nightjars, some species produce slowly or rapidly repeated churring sounds, in some cases rather toadlike, while others produce a warbled or whistled song, not at all disagreeable. Whistled phrases may be repeated hundreds of times without a break. Many different types of calls are used in different contexts, but the primary song, different for each species, appears to be the most important mechanism for species recognition. Besides vocal sounds, many species produce mechanical sounds in displays, such as claps and whirrs of the wings. In some species the vocalizations are given primarily in flight, in others primarily when perched; still others have different calls or songs when flying than when perched.

Echolocation by the oilbird

Oilbirds produce a variety of sounds, from clucking calls, apparently used to maintain contact, to harsh snarls and screams, when excited. They seem to have no primary song, which may be superfluous to their gregarious and specialized mode of life. Of greatest interest are the rapidly pulsated clicks that are emitted in shorter or longer bursts and that D.R. Griffin has found are used by the oilbird for navigation in total darkness on the principle of echolocation. These clicks are in the range of 7,000 cycles per second and thus audible to humans, in contrast to the ultrasonic pulsations used by bats in their system of echolocation. So far as is known, the only other birds to employ echolocation are some of the cave-swiftlets (*Collocalia* species). To what extent oilbirds use echolocation outside the caves is not known. As do their relatives in the order, they have large, sensitive eyes and probably navigate visually outside their caves.

Reproduction. Nesting habits are diverse among caprimulgiform families. Nightjars make no nest, depositing their eggs directly on the ground or on the leaf-covered floor of woodland. Occasionally they use a slight natural depression or scratch out a place among the debris. Populations of a few species (notably the North American nighthawks) have adapted to urban life and nest on flat gravel-covered rooftops. The one or two slightly glossy eggs may be cryptically coloured (profusely marbled with pink, buff, brown, or gray) or contrast markedly with their surroundings (white to buffy, usually with slight lines or blotches of colour). Incubation lasts about 19 days and may be by both adults or only by the female, there being considerable variation in this behaviour even within some species.

In contrast to the simplified nesting of nightjars, frogmouths construct nests on the horizontal forks of trees. In the genus *Podargus* the nest is of twigs and other plant matter and the two or three eggs are white; in *Batrachostomus* the nest is a pad of the birds' own down, bound and camouflaged externally with cobwebs and lichens, one white egg being laid. Both sexes are believed to incubate, the period being about 30 days. Owlet frogmouths nest mostly in hollow trees but also in tunnels in banks. They lay three or four white eggs, usually on a mat of leaves or fur. The incubation period is as yet unknown. The potoos are the most highly restricted nesters, for they must find a branch or stub with a suitable depression or crevice of just the right size to accommodate the single egg they lay. The egg is flat white, marked with brown and gray, and is incubated by both parents for 30 to 35 days.

On ledges in the caves they inhabit, the oilbirds build up a shallow nest rim of viscous, regurgitated fruit matter, which sets to a firm structure. As the accumulation of seeds raises the level inside the nest, the rim is constantly added to, so as to maintain a slight depression. The nests are used year after year and gradually grow to low mounds. Inevitably some seeds and the feces of very young nestlings contribute to the fabric of the rim. Two to four eggs are laid (white, subelliptical, slightly rough surface); both sexes incubate, the period being about 33 days.

Behaviour of the nestling

The chicks of nightjars are semiprecocial: they are down covered (buffy to brownish, plain or mottled, highly cryptic), fed by both parents, capable of opening their eyes on the first day (normally keeping them closed for several days), capable of walking on first day, and able to hop or run very well by third or fourth day. They usually run with wings raised, resembling young ducklings in gait. From the beginning chicks may solicit feeding by walking to the front of the adult, reaching up, and pecking at its bill. During the first days they are fed a regurgitated whitish viscous substance but later only partially digested insects. As early as the second day, although sometimes not until a week later, the chicks are led to different resting sites, at first only a metre or so at a time, but gradually much greater distances. Some of this moving may result from the chicks' own wanderings, for they are often restless and leave the covering adult.

In many cases both parents brood the chicks, in others only the female does so, and rarely the male assumes the primary role. By about two weeks the young are well feathered and too large for successful brooding; they rest at first near a parent but later quite alone.

Undisturbed young birds fly voluntarily at from 22 to 28 days; disturbed young may fly much earlier. About the time they are able to fly they begin a certain amount of feeding on their own but are still largely dependent up to 40 to 50 days. When two young hatch, one of them frequently develops more slowly and most often disappears during the second week. If it survives, however, it may require up to 40 days to reach the state of voluntary flight.

The nestlings of frogmouths, potoos, and owlet frogmouths are semidependent, covered with whitish down, fed by both parents, and remain in the nest until fledged. One observer noted that a young common potoo began wandering over the boughs of the nest tree at about four weeks. This same nestling first made trial flights at 47 days and finally left the nest tree when 50 days old. Other reports indicate the nestling period of potoos to be of 40–45 days, that of frogmouth to be about 30 days. The young are brooded only during the first half of this period, by which time young potoos have attained the juvenile plumage (whitish mottled with brownish) and are already accomplished in assuming the "broken branch" posture of adults.

On the whole, the caprimulgiform birds, like most birds, breed solely or mostly at the time of maximum abundance of food supply. Oilbirds may nest to some extent throughout the year, but the maximum nesting takes place during that time of year when the greatest number of appropriate trees are fruiting. Baby oilbirds are helpless (altricial), hatched blind and virtually naked, with only a little sparse down. They have an exceedingly slow rate of development. A thicker coat of down replaces the initial one after three weeks, and the adult plumage begins to appear at five weeks, there being no special juvenile plumage. The nestlings are apparently fed semidigested fruit pulp during the first days. At about two weeks there is a gradual change to whole fruits, regurgitated by the adults into the mouths of the young. As a result of the oil-rich fruit diet, young oilbirds build up extraordinary deposits of fat, which has led to economic exploitation. The fat can be boiled down to a high quality, durable oil, suitable for cooking and lighting. For this reason oilbirds were once subject to heavy predation by their human neighbours, a practice that continues even today, on a much reduced scale because of a general trend toward protection. The fat buildup reaches its maximum at about 70 days, at which time a young oilbird may weigh half again as much as an adult. By this time the plumage is well developed but not fully grown. The weight decreases as the feathers grow, until at an age of 100–120 days the young leave the nest, flying well and looking like adults. Both parents participate in feeding and brooding, the latter ceasing after about four weeks.

Migration. A number of tropical caprimulgiforms are sedentary, but the widespread cosmopolitan nightjars exhibit all degrees of migration. These migrations may be short, even local, or quite long. Some populations of the common nighthawk (*Chordeiles minor*) of North America and the European nightjar (*Caprimulgus europaeus*) migrate south to Argentina and South Africa, respectively. Actually, transequatorial migrations in both directions occur, as birds take advantage of the general alternation of seasons, and hence of food supply, on opposite sides of the Equator.

Lowering of body temperature in torpor

Hibernation. Some nightjars are able to cope with temporary food shortages by entering into periods of torpidity, a faculty they share with some swifts, hummingbirds, and a few others. The only known instances of apparently regular, prolonged annual hibernation in these birds, however, are reported for the poorwill *Phalaenoptilus nuttallii,* a nightjar. A banded individual was observed hibernating in the same small hollow in a rock during several successive winters. The bird was inert, with respiration and heart rates reduced to almost immeasurable levels and body temperature about 22° C below normal. Subsequent laboratory experiments have shown that several species of nightjar have the ability to relax into a torpid state under abnormal conditions. Such an efficient means of conserving energy may be more generally possessed than is presently realized throughout a family whose habits largely limit the birds to two brief periods of feeding a day, causing them to be especially vulnerable to involuntary fasting if inclement weather should prevent feeding during those periods.

FORM AND FUNCTION

Members of this order are easily recognized by their extremely wide mouths, large eyes, short legs and small, weak feet, and, except in the oilbird, soft plumage, in which browns and grays predominate.

The size of the gape is astonishing. When the bird opens its mouth, the opening seems to span the entire head, which is nearly the case. The beak, always somewhat hooked, is large and horny in the oilbird and frogmouths but reduced to a small projection in the remaining families. In body proportions, caprimulgiforms often appear chunky on account of the fluffiness of their plumage, but their actual bodies are proportionately no stouter than most songbirds. The tail is of medium length, except in males of those species with ornamental tail features. The wings are medium to long, and rounded in most species, pointed in a few, especially the nighthawks.

The plumage of all forms presents unexcelled examples of natural camouflage. Coloured in rufescent to ochreous browns, grays, white, and black, the species are variously patterned in greatest accord with their normal surroundings during daytime rest. Those nightjars that roost inside woodlands are streaked and spotted in a way resembling fallen leaves and other detritus on the forest floor. Those that dwell on gravelly terrain are speckled or otherwise patterned. The latter often contrast rather than blend with the soil, appearing like one of the many stones scattered about. The owlet frogmouths appear as clusters of dead leaves. Potoos and frogmouths are streaked and mottled like bark so that in the daytime alarm posture they appear most effectively as dead stubs.

Red and gray colour phases

Caprimulgiform species are frequently dichromatic, having grayish and reddish phases. These appear to be randomly distributed in some cases and sex-linked in others, there being several species of nightjars in which males are more grayish, females more reddish.

Except for possible freaks, the iris colour is brown in nearly all members of the order. Yellow occurs in the irises of some potoos, and yellow, orange, and ruby in frogmouths. Contrary to some reports, oilbirds do not have blue irises nor are any nightjars known to have yellow eyes.

CLASSIFICATION

Distinguishing taxonomic characters. The most important characters used to define the caprimulgiform birds are the type of plumage, structure of the feet and legs, form and structure of the bill, palatal structure, arrangement of the pelvic muscles and flexor tendons, shape of dorsal vertebrae, meristic characters such as the number of primary and secondary feathers and rectrices, presence or absence of intestinal ceca and of an oil gland, location of syrinx, presence and type of rictal bristles, powder down patches, carotid artery relationships, and a few other anatomical details. Variations in some of these same features characterize the different subgroups.

Annotated classification. The arrangement presented is the same as that proposed by A. Wetmore (1918) in a modification of R. Ridgway (1914). The group was then considered a suborder Nycticoraciae of the order Coraciiformes. The subsequent shift to ordinal level with the name Caprimulgiformes has long been accepted.

ORDER CAPRIMULGIFORMES

Soft plumaged (except Steatornithes), cryptically patterned birds with relatively weak anisodactyle feet and very short tarsi; deeply cleft gape (except Steatornithes); pelvic muscle formula AXY (XY in Steatornithes); flexor tendons fused (synpelmous); 10 primaries; 11–13 secondaries; 10 rectrices; aftershaft small but present; 2 carotid arteries and oil gland present (except in *Podargus* and *Nyctibius*).

Suborder Steatornithes

Dorsal vertebrae opisthocoelous (concave behind); gape not exceptionally deeply cleft; rostrum movably articulated with skull; plumage firm.

Family Steatornithidae (oilbird)

No fossil record. Locally distributed in Guyana, Venezuela, Colombia, Ecuador, Peru, and Trinidad. Desmognathous palate (maxillopalatine bones fused), palatines narrow and not expanded posteriorly; large, strong bill with hard rhamphoteca (horny covering) and subterminal tooth, surrounded by long vibrissae; bronchial syrinx; 15 cervical vertebrae. One species, length 40 centimetres.

Suborder Caprimulgi

Dorsal vertebrae heterocoelous (saddle-shaped); deeply cleft gape; rostrum fixed; plumage soft.

Family Podargidae (frogmouths)

No fossil record. Confined to Australasian (except New Zealand) and southern Oriental regions (including extreme southern India). Desmognathous palate; palatines broad throughout, slightly expanded posteriorly; wide, strong bill with hard rhamphotheca; bronchial syrinx; 13 cervical vertebrae; well-developed, powder-down tufts on either side of rump; oil gland absent in *Podargus,* very small in *Batrachostomus;* one carotid artery (left). Twelve species, length 20–40 centimetres.

Family Nyctibiidae (potoos)

Upper Pleistocene to present. Neotropical region. Schizognathous palate (small vomers and separate maxillopalatines); palatines narrow anteriorly, greatly expanded posteriorly; small, weak bill but with prominent horny angular projection on maxillary tomium midway between tip and rictus and tomium very broad, horny, and strongly convex from that point to rictus; extremely short tarsi; toes unusually wide basally, forming a broad flattened sole; tracheobronchial syrinx; 14 cervical vertebrae; large powder down patches on sides and breast; no oil gland; one carotid artery (left). Five to seven species, length 20–55 centimetres.

Family Aegothelidae (owlet frogmouths)

No fossil record. Australasian region (except New Zealand). Desmognathous palate; bill similar to Podargidae but shorter, weaker, and largely hidden by forehead feathering; bronchial syrinx; unique in the order in lacking ceca. Eight species; length 16–25 centimetres.

Family Caprimulgidae (nighthawks, nightjars)

Earliest probable fossil Oligocene, earliest modern fossil *Caprimulgus* in Pliocene (Europe). Worldwide in tropical and temperate zones; absent from northernmost Eurasia and America, southernmost South America, New Zealand, and some oceanic islands; subfamily Chordeilinae (nighthawks) restricted to New World. Schizognathous palate (except *Chordeiles* and perhaps other Chordeilinae); palatines narrow anteriorly, greatly expanded posteriorly; small, weak bill; well-developed rictal bristles (except *Eurostopodus* and Chordeilinae); small, weak feet; lateral toes much shorter than middle toe, fourth toe with only four phalanges (segments; five is normal for this order), middle claw with inner edge pectinated; hallux very short and directed inward; tracheobronchial syrinx; 14 cervical vertebrae. Seventy to eighty species; 14–40 centimetres (without ornamental feathers).

Critical appraisal. The major divisions of the caprimulgiform birds are clearly defined and their composition remarkably homogeneous so that there are no apparent taxonomic problems above the level of genus. The rather aberrant oilbird has traits similar to owls, but evidence places it in this order. Studies of the egg-white proteins by C.G. Sibley show that it is clearly caprimulgiform. These studies indicate a close relationship between the nightjars and the owls but do not support one to the swifts. Within the order the main taxonomic problems relate to re-evaluation at the generic level and clarification of species limits. These are likely to be settled only with the aid of field studies utilizing bioacoustical methods along with other disciplines. (P.A.S.)

Apodiformes (swifts, hummingbirds)

The order Apodiformes contains two groups of birds that are very different from one another in general appearance and way of life: the swifts and the hummingbirds. The two groups, considered suborders, are Apodi, which contains two families, Hemiprocnidae for the tree swifts (also called crested swifts) and Apodidae for the true swifts; and Trochili, which contains the single large family Trochilidae for the hummingbirds. The most obvious common characteristic of the two groups is a superb mastery of the air; both swifts and hummingbirds are agile fliers and share, in this connection, some peculiarities of wing structure. Some ornithologists believe that these anatomical resemblances are convergent (the result of the same type of natural selection acting on unrelated organisms) and that the true relationships of hummingbirds and swifts lie with other groups, not with each other.

GENERAL FEATURES

Importance to man. Man has long been fascinated by hummingbirds. From the time the first specimens arrived in Europe from the newly discovered American tropics, the brilliant, iridescent colours and diminutive size of hummingbirds have made them favourites of both scientists and the public. Stuffed hummingbirds were in great demand for display in homes and museums and for decorating ladies' hats. South American Indians brought hummingbirds from hundreds of miles around to great market centres such as Bogotá, Colombia, and Cayenne, French Guiana. Even today, some species and subspecies of hummingbirds are known only from these "trade skins," and their actual range has never been discovered by ornithologists. The commerce in hummingbird skins has almost completely vanished, and collectors now capture hummingbirds alive for exhibition in zoos and aviaries. Although some species are surprisingly hardy in captivity, the proper care of hummingbirds is not simple, and few of the birds that leave the jungle survive.

Bird's nest soup

In most parts of the world, swifts, little known to the average person, are often confused with swallows, which, although unrelated, are generally similar to swifts in size, proportions, and aerial habits. In Southeast Asia, however, one group of swifts is of major economic importance. The swiftlets of the genus *Collocalia* comprise a group of species, most of which live in caves. They build nests composed of varying amounts of plant and animal substances (*e.g.*, leaves, moss, hair, feathers) held together and fastened to the cave wall with a mucilaginous secretion of the salivary glands. The nest of one species, the edible-nest swiftlet (*C. fuciphaga*), is composed almost entirely of concentric layers of this salivary cement. These nests and, to a lesser extent, those of some other swiftlets are gathered commercially in the East Indies and form the base for the famous bird's-nest soup of the Orient.

Size range and diversity of structure. The smallest of living birds are included among the Trochilidae, the smallest known species being the bee hummingbird (*Mellisuga helenae*) of Cuba, which is barely 62 millimetres (two and one-half inches) from bill tip to tail tip. Even the largest member of the family, the giant hummingbird (*Patagona gigas*) of western South America, has a body only about the size of that of a large sparrow, although the bill and tail extend this length by about 35 and 85 millimetres (one and one-half and three and one-half inches), respectively. There is relatively little structural diversity among hummingbirds except in the size and shape of the bill and the ornamental feathers. The bill may be very short or, as in the sword-billed hummingbird (*Ensifera ensifera*) of the Andes, may exceed the rest of the bird in length. The bill may also be straight or curved to varying degrees, usually downward, but, in a few species, upward. In most hummingbirds, the tail feathers are relatively unspecialized; but in some others they are greatly elongated, either straight or curved, in others wirelike with spatulate tips. Brilliantly iridescent feathers are found somewhere on the body of most hummingbirds, often with an especially colourful gorget on the throat; a few genera have dull coloration. Sexual dimorphism (difference between sexes) varies, with females being either indistinguishable from males or less colourful or ornate.

Variety of swifts

The true swifts have a somewhat greater size range, from such tiny species as the five-gram (0.2-ounce) pigmy swiftlet (*Collocalia troglodytes*) of the Philippines to some of the large and powerful members of the Old World genus *Apus,* 30 times heavier. Beyond the size differences, the most obvious morphological variation among swifts is in the conformation of the tail. The spine-tailed swifts, a group that includes the familiar American chimney swift (*Chaetura pelagica*), have the central shaft of the tail feathers extended beyond the vanes (flat, expanded part) into pointed spines. In other swifts, with less specialized tail feathers, the tail itself may be square ended, moderately forked, or deeply forked (swallow-tailed). Less conspicuous structural variations that are used in swift taxonomy include the position of the toes and the amount of feathering on the legs. The four species of tree swifts that comprise the family Hemiprocnidae are rather uniform, differing mainly in size, proportions, and plumage pattern.

Hummingbird distribution

Distribution and abundance. Although there are unrelated Old World birds, such as the sunbirds (Nectariniidae), that superficially resemble hummingbirds, the family Trochilidae is confined entirely to the Americas. Of approximately 320 known species, most are tropical, and the greatest numbers of these are found in northwestern South America (Colombia and Ecuador). Several species have become adapted to temperate conditions, some by having extended their distributions north or south of the tropics, others by their occupation of high-altitude habitats in South and Central America. In North America only one species, the ruby-throated hummingbird (*Archilochus colubris*), is normally found east of the Great Plains. About

Drawing by Christian D. Olsen

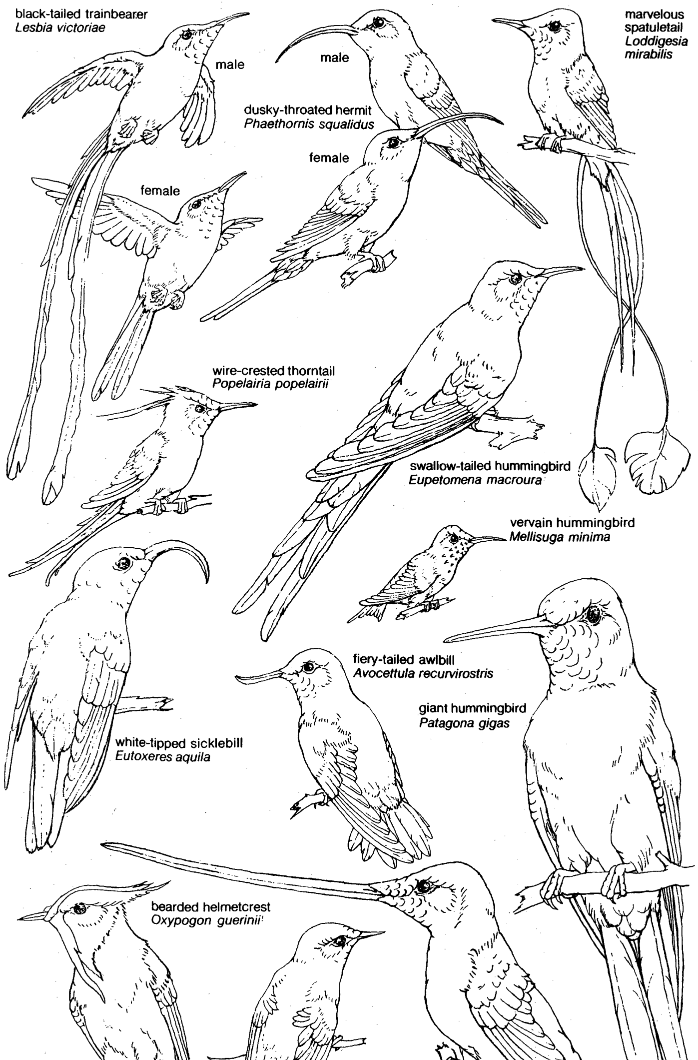


Figure 39: Body plans of hummingbirds.

a dozen occur regularly in the western United States; one, the rufous hummingbird (*Selasphorus rufus*), reaches southern Alaska.

The family Hemiprocnidae is confined to Asia, from India south and east through the East Indies, with one species reaching as far as the Solomon Islands. The Apodidae are much more widely distributed, occurring in both the Old and New Worlds. There are approximately 70 to 80 species, but the features that distinguish closely related species are often so subtle that there is little agreement about swift classification at either the genus or species level. Virtually nothing is known about some species other than their existence as evidenced by museum specimens; such birds may be rare, shy, or both. Others, such as the chimney swift, the European swift (*Apus apus*), and the house swift (*Apus affinis*) of India, are familiar inhabitants of towns and have been extensively studied by biologists.

NATURAL HISTORY

Behaviour and locomotion. In such overwhelmingly aerial birds as swifts and hummingbirds, locomotion is an inseparable part of behaviour. Although both groups depend on their flying ability to obtain food, the actual feeding methods are quite different. Swifts feed exclusively on flying insects, sharing with the unrelated swallows several morphological adaptations for this habit (*e.g.,* the short, weak bill, large mouth, and long, narrow wings). The feet of the true swifts are much weaker in proportion to body size than are those of swallows or hummingbirds, and the former are unable to perch on twigs, wires, or other similar landing places. In their normal resting posture the true swifts cling to vertical surfaces except when sitting on the nest. The tree swifts are somewhat less specialized and are fully capable of perching; in fact, their usual foraging method involves launching after flying prey from the vantage point of a high perch. Swifts drink by swooping down at the surface of a body of water, and nesting material may be picked in midair, as in the case of drifting feathers or plant material, or snatched in full flight, as are dead twigs and grass stems. Most courtship displays that have been described for swifts are wholly aerial, and copulation is known to take place in full flight.

Some species of swiftlets (*Collocalia*) that nest in total darkness deep in caves have the ability to orient themselves in flight by echolocation (reflection of sound waves back to the emitter), a phenomenon otherwise known, among birds, only in the oilbird (*Steatornis caripensis,* of the order Caprimulgiformes). The swiftlets emit a rattling call consisting of a succession of clicks, which, unlike the calls used by bats for the same purpose, are fully audible to the human ear. A few species of swiftlets lack this ability and nest only where some light penetrates the caves.

Feeding behaviour of hummingbirds

Hummingbirds are less constantly aerial than swifts and often spend long periods quietly perched. In foraging, however, most species depend upon their unique ability to dash from place to place and then to hover, motionless except for beating wings and darting tongue. They can even fly backward or upside down for short distances. The basic food is flower nectar, and the shapes of hummingbird bills are sometimes closely correlated with the shapes of the flowers from which they feed. The long, tubular tongue serves also for sucking up dew or rainwater as needed. Hummingbirds also eat small insects and spiders, which usually appear to be captured with the bill rather than with the tongue. Although hummingbirds perch, they share with swifts the lack of walking ability or other nonflying locomotion. They may sometimes sidle on a perch, just as swifts may crawl up a wall, but normal locomotion is by flying, even for short distances.

Prenesting, or courtship, behaviour is varied in hummingbirds; territoriality is completely absent in some species; others fiercely defend nesting or feeding territories. Hummingbirds in general tend to be aggressive and may attack vastly larger potential predators, such as crows and hawks. Leks, areas of communal male displays, have been described in some tropical species. Most displays are aerial, and some are spectacular; in other species, such as the plainly coloured hermits of the genus *Phaethornis,* singing forms an important part of display behaviour.

Reproduction and life cycle. The nests of tree swifts are tiny cups, barely large enough to hold the single egg. They are placed on small limbs, high up in trees, where they are subject to being moved about by the wind. The egg is glued to the bottom of the nest with the sticky saliva of the parent; this prevents the egg from being turned during incubation (as is done by virtually all other birds), and it is thought that the waving of the nest branch in the wind may in some way act as a substitute, moving the egg's contents within the shell. One of the true swifts, the palm swift (*Cypsiurus parvus*) of the Old World tropics, also glues its egg into the nest, but the nest is very different. This species utilizes feathers and plant material and attaches the nest, little more than a flat pad, to the underside of a hanging palm leaf, where it, too, is subject to agitation by the wind.

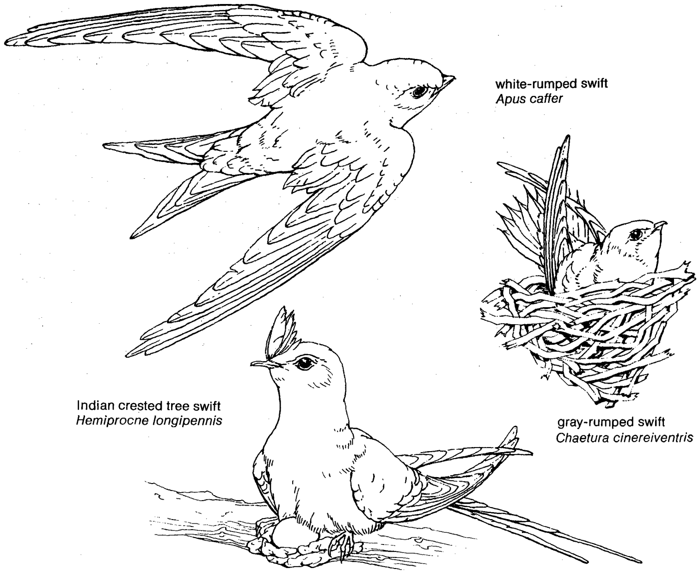

Drawing by Christian D. Olsen

Figure 40: Body plans of swifts.

Variations in swift nesting habits

For such a relatively uniform group in a structural sense, the true swifts show a surprising variation in nesting habits. The New World palm swifts (*Tachornis*), like those of the Old World, place their nests under palm leaves, but the nest itself is more elaborate, being a sack entered from below through a tubular tunnel. Most other swifts nest either inside a crevice or hole (*e.g.,* in the ground, in a tree, between rocks), where they build a shallow cup of feathers and plant material as a flooring, or inside hollow trees, caves, or such man-made structures as buildings, wells, and tunnels. The nests may be circular cups placed on ledges or semicircular brackets attached to the walls. The bracket nests are fastened to vertical surfaces with a sticky secretion from the large salivary glands; in the swiftlet genus *Collocalia,* this substance may comprise most or all of the nest material itself. One species, the white-naped swift of Mexico, builds no nest at all but lays its eggs in a depression on bare sand on ledges deep inside caves.

Clutches in the true swifts vary from one to about six white eggs, with the higher numbers being found among some of the more northern species of the genera *Chaetura* and *Aeronautes.* Incubation is by both sexes. The young are hatched completely naked. Young swifts are left unattended for longer periods than is true of most altricial (helpless) nestlings; the adults spend a substantial amount of time foraging for the relatively small insects on which they feed, in order to capture a sufficient number to warrant a trip back to the nest. In bad weather, when few insects are flying, the young may be left unbrooded for hours. In some birds, this would normally lead to chilling and, perhaps, death, but young swifts maintain an essentially poikilothermic (cold-blooded) condition for the first few weeks of life, the body temperature dropping to near that of the environment without harming the young bird. In one genus, *Cypseloides,* the young develop a coat of downlike semiplumes (fluffy feathers) prior to the appearance of the juvenile pennaceous (smooth) contour feathers, and this fluffy coat also appears to aid in thermoregulation.

Many species of swifts are colonial nesters, and even those that are relatively solitary during the breeding season appear to be highly gregarious at other times of year. Night roosts of some species may number hundreds of individuals, and foraging flocks may include several species. The northernmost species, such as the European swift and the chimney swift, are long-distance migrants. The winter home of the latter species was a mystery for many years, until Indians in the upper Amazonian forests of Peru were found to have in their possession numbered aluminum bands that had been placed on the legs of chimney swifts by the United States Bureau of Biological Survey. Migration is also known to occur in some of the tropical species, notably in Africa.

The Indian population of the house swift has two breeding seasons per year and is one of the few birds in the world in which this phenomenon has been demonstrated. Adult males are in full breeding condition in late January and again in May and June; eggs are laid in January and February and again in June to September. The molt cycle, however, appears to be independent of the breeding cycle, and the flight feathers are renewed only once a year.

Nest design among hummingbirds

Unlike swifts, hummingbirds are largely polygamous. Males may build nests, but they are not used by the females. Most hummingbird nests, saddled on a branch, are composed of fine vegetable fibres, spider webs, fragments of bark, and lichens. Among the most interesting exceptions are the nests constructed by the hermits. They are long, slender, woven baskets of plant fibres and down, suspended from a strip of vegetation attached at only one side of the nest. The weight of the incubating female and, later, the weight of the young is counterbalanced by pebbles or clods of hard earth at the bottom of the dangling nest. The usual hummingbird clutch is two white eggs, and the female alone incubates and cares for the young in most species, although this is difficult to determine in the species that are not sexually dimorphic. In the sexually dimorphic species, the juveniles usually resemble the adult females, although some genera have a distinctive juvenile plumage that resembles neither parent. The young are fed by the parent thrusting its bill deep within the young bird's throat and regurgitating food.

Like the northern swifts, the hummingbirds that breed farthest north are the most highly migratory. The ruby-throated hummingbird was known to make annual nonstop flights across the Gulf of Mexico long before physiologists understood avian metabolism well enough to determine how this could be possible for such a tiny bird. Migration takes place among some Southern Hemisphere hummingbirds, such as the giant hummingbird and the green-backed fire crown (*Sephanoides sephanoides*); the latter reaches Tierra del Fuego and is thus the southernmost hummingbird, but there is little information about the nature and extent of these movements.

Ecology. The distribution of both swifts and hummingbirds is strongly influenced by the specialized nature of their foods. Swifts are found virtually worldwide but are limited to areas where flying insects are found in sufficient numbers to sustain adults and young. Food of this kind is presumably more abundant in the tropics, and the family is primarily a tropical one; it is chiefly in the tropics that sympatric species (*i.e.,* those whose ranges overlap), sometimes closely related, may be found. Although there may be some altitudinal and other ecological separation of sympatric species, virtually nothing is known about possible partition of environmental resources among swifts; as many as three closely similar species of *Chaetura* can be seen foraging together in Trinidad. Swifts also forage with swallows.

Hummingbirds are more specialized in their feeding habits, in that their bill shapes are, in many cases, correlated with the shapes of the most commonly utilized flowers. Because hummingbirds are restricted almost entirely to feeding at flowers, their distribution is affected by flowering periods, and the virtually seasonless tropics, where flowers of some kind may be found the year round, are, not surprisingly, the home of the most species. Despite the degree of specialization found, a flowering tree in the tropics may attract six species of hummingbirds simultaneously. In some cases feeding areas are strongly defended territories; in others the birds feed side by side, apparently ignoring each other completely.

Lowered metabolic levels in resting birds

Torpor. Swifts and hummingbirds are among the few groups of birds that have been shown to be capable of energy conservation through a reduction of their body temperature and entry into a torpid (sluggish) condition. The extremely high metabolic rate of active hummingbirds suggests that the value of torpor to the birds lies in energy conservation while roosting at night. In the swifts, whose normal metabolic rate approaches that of other birds, the lowered body temperature is an energy-conserving mechanism during what may sometimes be long periods of fasting, at times when foraging for flying insects is not possible.

Evolution, paleontology, and classification. *Fossil history.* As might be expected of such fragile birds, no fossil species of hummingbirds are known, although remains of Pleistocene origin (less than 2,000,000 years old) have been reported for two living species. Five fossil swifts have been described; two from the lower Miocene age (about 20,000,000 years ago) and one from the upper Miocene (about 10,000,000 years ago) have been assigned to modern genera. The extinct genus *Cypselavus* is known from two fossil deposits in France: one species from the lower Miocene, the other from the upper Eocene or lower Oligocene (35,000,000–40,000,000 years ago). Their identification as swifts does not seem to be questioned, but they throw no light on the origins of the family. The fossil family Aegialornithidae, sometimes provisionally assigned to the Apodiformes, is now usually considered to belong to the order Caprimulgiformes.

Classification. As presently constituted, the subdivisions of the order Apodiformes are summarized below.

ORDER APODIFORMES

Characterized primarily by the shape and proportions of the wing. The humerus (bone of the upper arm) is short, the hand is strikingly long, and the primary flight feathers (attached to the hand) are much longer than the secondaries (attached to the forearm). Other anatomical characteristics shared by swifts and hummingbirds include the holorhinal (the posterior border of the internal openings rounded) and impervious (without external openings) condition of the nostrils; single (left) carotid artery (except in swifts of the genus *Cypseloides,* which have two); sternum without notches; and oil gland unfeathered.

Suborder Apodi

Bill short and gape (mouth) deeply cleft; tongue short; salivary glands large; crop absent; nostrils without opercula (coverings); 8–11 secondaries; 6–7 pairs of ribs.

Family Hemiprocnidae (tree swifts)

Hallux (hind toe) directed backward, not reversible, foot capable of perching; no claw on manus (hand). Hatched with natal down; a distinct juvenile plumage unlike that of adults; adult plumage with crests or plumes on head. One genus, 3–4 species.

Family Apodidae (true swifts)

Approximately 70–80 species; species limits, especially in *Collocalia,* highly controversial. Hallux lateral or nearly so, reversible in some; foot incapable of true perching; claw present on hand. Young hatched naked; juvenile little different from adult in plumage; no crests or plumes. One fossil and 8 to 14 Recent genera; 70–80 species.

Suborder Trochili

Bill slender, usually long, gape not deeply cleft; tongue long, tubular, and extensible; crop present in nestling; nostrils with opercula; 6–7 secondaries; 8 pairs of ribs.

Family Trochilidae (hummingbirds)

Characters of suborder. More than 120 genera usually recognized, but a further study will probably result in reduction. Approximately 320 species, nearly 30 of which are known only from trade skins; some may be hybrids.

Critical appraisal. The most fundamental question in the classification of the Apodiformes is whether the swifts and hummingbirds are, in fact, more closely related to each other than to any other birds. Some ornithologists believe that the resemblances in wing structure are convergent and that the other shared characters are relatively widespread among several orders of birds and of little significance in indicating relationships. These authorities argue that the differences between swifts and hummingbirds are so fundamental that the two groups could not have shared a common ancestor recently enough to warrant placement in the same order. Ornithologists holding this opinion would recognize a separate order, Trochiliformes, for the hummingbirds. The Caprimulgiformes (nightjars and allies) are usually considered to be the closest relatives of the swifts, but the placement of the hummingbirds is more difficult. Their closest living relatives, if not the swifts, must almost certainly be among the group of orders usually placed nearest to the Passeriformes (perching birds).

The tree swifts have sometimes been subordinated to the level of a subfamily of the Apodidae. Their many relatively nonspecialized anatomical characters indicate that, although indubitably swifts, they merit separate family status. Two subfamilies, usually Apodinae and Chaeturinae, are customarily recognized in the Apodidae but probably represent an exaggerated dichotomy. The African species *Schoudetenapus myoptilus* has been shown to combine certain characteristics of both subfamilies and may be a relict of an ancient and relatively primitive group of Apodidae. Additional studies may enable ornithologists to redefine subfamilies of swifts.

Problems of classifying swifts

The classification of swifts at the genus and species level is also fraught with controversy. Part of this represents the traditional battle between advocates of large, inclusive genera and those who prefer to use generic names to call attention to differences; in other classifications, one difficult species is transferred between what are otherwise generally accepted genera. Species limits are especially difficult to determine among the swiftlets (*Collocalia*); there is absolutely no agreement about how many species exist nor what names they should bear. The characters that have long been used to define species of *Collocalia* are mixed, and some have proved to be of no taxonomic importance. Important morphological characters have been overlooked; one swiftlet, first described as a subspecies, was found 26 years later to be the only three-toed form in the genus and was quite properly given species status. The type of nest construction and the presence or absence of the ability to navigate with echolocation appear to be fundamental characters within *Collocalia,* but this information is not available for many forms, and the taxonomy of the genus will long be in a state of flux.

Among the hummingbirds there are obvious groupings of related genera, but no discontinuities are marked enough to warrant the recognition of subfamilies. There are generally conceded to be too many hummingbird genera that are based entirely on characters found only in males, involving ornamentation used in species-specific displays. As yet, however, there has been no research that will permit the redefinition of hummingbird genera. (K.C.P.)

Coraciiformes (kingfishers and allies)

The birds of the order Coraciiformes include the kingfishers, todies, motmots, bee-eaters, rollers, hoopoes, and hornbills and are collectively referred to as roller-like birds. Among the members of the order that have attracted man's special attention are certain kingfishers that plunge headfirst into water for fish and are associated with classical mythology. According to the ancient Greeks, Ceyx and his wife Alcyone were shipwrecked at Delphi and changed into kingfishers. The Chinese used the shining blue feathers of some types of kingfishers to decorate picture screens. Bee-eaters (Meropidae) have been accused of preying on commercially valuable honeybees, and the North American belted kingfisher (*megaceryle alcyon*) is sometimes considered a pest at fish hatcheries because it preys on young game fish. The kookaburra (*Dacelo novaeguineae*) has a loud, laughing or braying voice that is commonly associated with the Australian outback, or back-country. To the biologist, the sealing of the female of certain species of hornbills in her nest during incubation and brooding is one of the most intriguing behavioral modifications among birds.

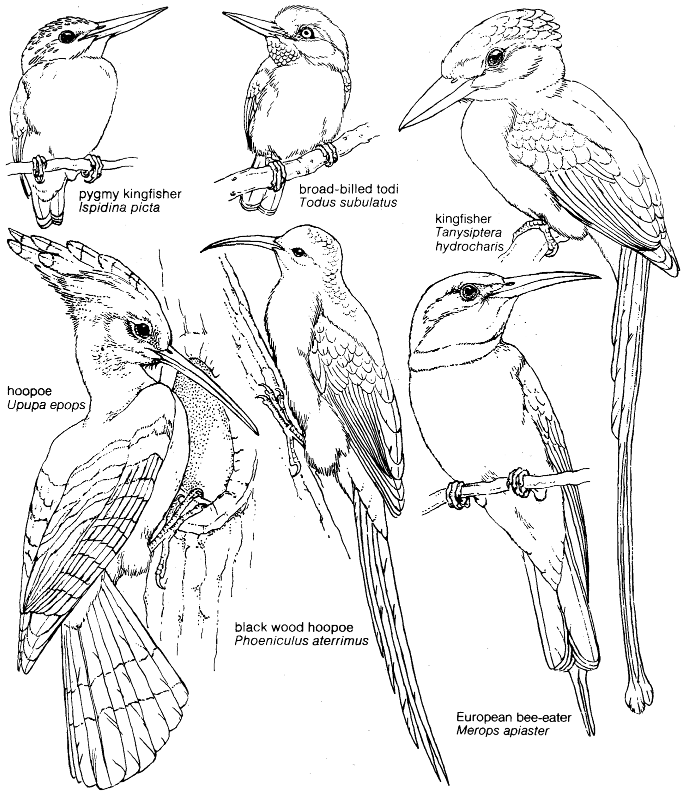


Figure 41: Body plans of smaller Coraciiformes.
Drawing by R. Keane

General features. The roller-like birds are active by day (diurnal) and range in length from the size of a small sparrow (about ten centimetres, or four inches) to over 100 centimetres (40 inches). They have compact bodies, short to moderately long necks, large heads, rather long bills, small feet, and ample wings. The tail varies from short to very long and may be forked, square, or graduated; the outer or central tail feathers are, in some species, pointed or spatulate at the tip. All of these birds regularly perch in trees, where some feed; others fly in search of food, and a few walk or hop on the ground. The group's food ranges from invertebrates (including insects) and small vertebrates to berries and fruit.

Distribution. Collectively, the ten families of the order are almost worldwide in temperate and tropical areas, with the greatest number and diversity in the warmer parts of the African, southern Asian, and Papuan areas. Many species are common and conspicuous, and a few tolerate human settlement, although their direct importance to man is minimal.

Families of coraciiform birds

Only the kingfishers (family Alcedinidae) are found in both Eastern and Western Hemispheres. The motmots (Momotidae) and todies (Todidae) are restricted to the New World tropics, the bee-eaters (Meropidae), rollers (Coraciidae), and the hoopoes (Upupidae) to temperate and tropical regions of the Old World, and the hornbills (Bucerotidae) to the tropics of Africa and Asia. The wood hoopoes (Phoeniculidae) are found only in tropical and subtropical Africa; the cuckoo roller (Leptosomatidae) and the ground rollers (Brachypteriacidae) are found only on Madagascar.

NATURAL HISTORY

Life history. Most species live permanently in one region, but temperate-zone species move nearer the tropics for the winter. In the Old World tropics that have a dry–wet seasonal change, local movements of bee-eaters sometimes take advantage of the related crops of insects. In East Africa, where there are two dry and two wet seasons, two periods of breeding may occur a year; however, the picture is not clear.

As a group, these birds are well endowed with voices, but their vocalizaions are usually referred to as calls rather than songs. Some are harsh; others are soft or are whis-

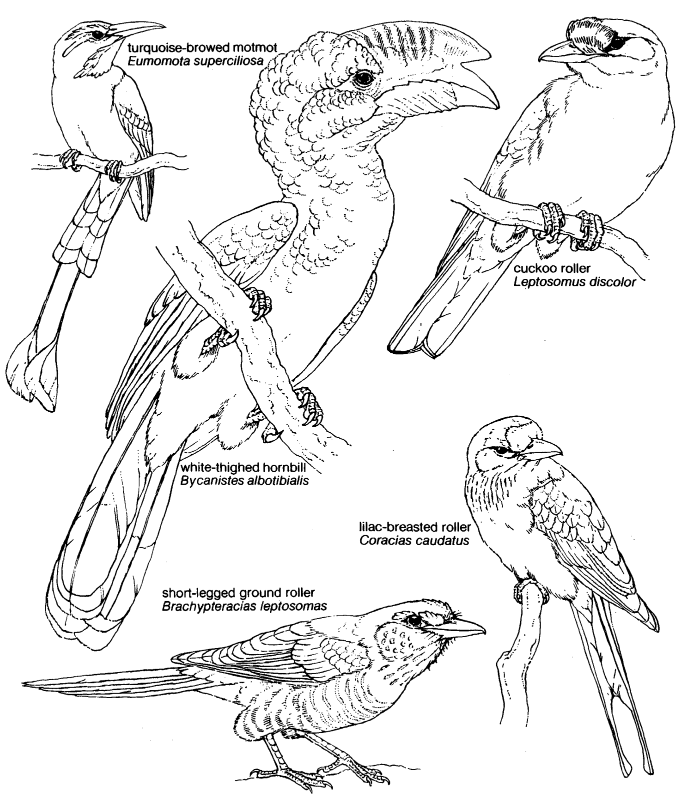


Figure 42: Body plans of larger Coraciiformes.
Drawing by R. Keane

tled or hissed. Some are given as single notes, others as series in a trill, a rattle, or a hooting, and still others as a cacophony or medley of notes. Some calls are given from perches; others are given on the wing. Certain utterances may be related to courtship and mating or to territory; others seem to be simply a part of the bird's general daily activity. Pairs are formed to the accompaniment of simple posturing displays and series of calls, ranging from harsh to soft. Certain rollers use tumbling display flights. Both sexes usually share in the nest duties.

Many roller-like birds are solitary in feeding and nesting, but some bee-eaters are gregarious and also nest in colonies. Hornbills and wood hoopoes move about in small parties most of the year, but they nest solitarily. The nest site is always in a cavity, which may be a hole or crevice in a tree, a bank, or a wall. The cavity may be among rocks or be a tunnel dug by the birds in the ground, or it can be an abandoned termite nest.

Walling in of the female hornbill

Normally, little or no nest material is added, except by the hoopoe (*Upupa epops*). In some (perhaps most) hornbills, however, the female enters the nest cavity before laying starts; the male brings mud and debris, which the female takes and plasters around the entrance until only a slit is left. The male passes food to the female through the slit until after the young are more or less grown. At that time she breaks out of the nest, and then both parents feed the young.

The members of the order lay two to nine eggs, with the tropical species laying the smaller clutches. Among the hornbills the larger species lay fewer eggs than do some smaller ones. The eggs are usually white but, in the hoopoe, may be olive, brownish, bluish, or greenish, and sometimes spotted. Incubation is performed by both male and female in kingfishers, todies, motmots, and bee-eaters and by the female alone in the hoopoe and hornbills when she is fed by the male. Incubation periods of 18 to 22 days have been recorded for some of the smaller members of the order.

The young are nidicolous (dependent upon the parents) and are naked, except for the hoopoe and some kingfishers that have varying amounts of down at birth. The recorded nestling period (25 to 28 days) of certain of the smaller species is only slightly longer than the incubation period, but large species have longer nestling periods, and the great hornbill (*Buceros bicornis*) has a total period (incubation to fledging) of three to four months. Except for the hoopoe and the walled-in hornbills, the young are fed by both parents. The male of these species brings food to the female, who passes it on to the young. The rate of feeding the young is apparently variable. In large species of hornbills, the male is recorded as making one trip per hour to the nest, during which he brings a gullet full of fruits and regurgitates them one at a time and passes them to the female. A much smaller species brings one item of animal food and holds it in his bill each trip and makes about six trips an hour.

Coraciiforms of several groups practice no nest sanitation, and regurgitated insects and other remains accumulate and are used to make a platform on which the young rest in later nest life. Hornbill nests are kept more or less clean, and, in one small species in which the female breaks out of the nest when the young are half grown, the young perform nest sanitation and also help to seal up the nest again after the female emerges.

When the young leave the nest, they are able to fly and usually are similar to, although sometimes duller or paler than, the adult in plumage. There is believed to be one molt a year, which occurs after the nesting season. In an unusual modification, the female of the African hornbill, walled in during incubation and throughout the nest life of the young, molts during this period, losing and renewing all of her flight feathers.

Locomotion and feeding. Coraciiform birds tend to perch in trees and shrubs when at rest. Some favour exposed perches on which they are conspicuous, others seek the protection of foliage or the shade of the forest. Lacking cryptic coloration—many are nearly uniform in colour or boldly patterned—they do not rely on concealment for protection. Their flight varies from weak and laboured to strong, well sustained, and direct. The flight of some, such as the rollers, is swift and graceful. Some species, such as the kingfishers, use little bipedal locomotion; others (*e.g.,* some hornbills) hop, walk, or scramble in the treetops, creep along branches (wood hoopoes), or walk or hop on the ground (hoopoe, other hornbills).

The food of the roller-like birds includes a wide variety of organisms. Among the animals taken are worms, snails, crustaceans, insects, fish, amphibians, reptiles, small birds, and mammals. Vegetable food consists chiefly of the fruits of trees, usually gathered in the trees but sometimes picked up on the ground.

Types of foraging behaviour

Some birds of this order seem to choose animal food more for its size than its type, and a single method of feeding tends to predominate in each family. Some families (Upupidae and Leptosomatidae) contain only one species; others are large and predictably diverse—*e.g.,* the hornbills (Bucerotidae), with 45 species, and the kingfishers (Alcedinidae), with about 90 species. Each family or group of families tends to have a characteristic pattern of feeding behaviour, and the foraging patterns fall into four categories, or feeding niches: (1) watchful waiting on a perch, (2) aerial—*i.e.,* spending much time on the wing, (3) searching on foot among branches of trees, and (4) walking on the ground.

The watchful waiters, the kingfishers, motmots, and todies, tend to sit quietly for long periods. When they see their prey on a leaf, a branch, the ground, or even in water, they fly out with swift, direct flight, seize the prey with the bill, and return to the perch. The todies catch more flying insects than do members of the other two families. A few kingfishers plunge headfirst into water from perches or from hovering flight, but these number only a few of the species-rich family Alcedinidae. The shovel-billed kingfisher (*Clytoceyx rex*) of New Guinea is partly terrestrial and is known to feed on beetles and earthworms; the latter are apparently dug from the soil of the forest floor with the bird's short, heavy bill. The ruddy kingfisher (*Halcyon coromanda*), widespread in Southeast Asia, eats many large land snails. It seizes a snail with its bill and beats it against a rock until the shell is broken and the meat can be extracted.

The term temperament, although tinged with human associations, seems applicable to certain traits that are common to many species within a family, as contrasted with members of another group. Kingfishers, motmots, and todies are stolid, phlegmatic birds that sit quietly for varying periods of time between sallies for food. Kingfishers often bob their heads and the forepart of their bodies when nervous or mildly alarmed; when startled into flight, some give sharp calls. Motmots have the habit of moving their tails from side to side.

Bee-eaters and rollers are aerial feeders but spend much time perched quietly, leaving their perches to make sallies for passing insects. Bee-eaters, especially, often spend long periods on the wing, gliding in circles while looking for insects, especially bees and wasps. Some forest bee-eaters perch in foliage and near flowers, securing their prey without flying. Rollers spend more time perched, but they too are graceful in flight and capture much of their food by hawking or by darting down to the ground. Members of both groups are often seen aloft, apparently not feeding but flying for diversion. The cuckoo roller (*Leptosomus discolor*) also flies above the forest canopy, but it is looking for large insects and small lizards in the outermost foliage. It may either seize them while on the wing or alight to capture them.

Feeding while clambering among the branches of trees is carried on by many of the larger hornbills and by the small wood hoopoes, but in quite different ways. The hornbills fly over or through the forest with strong, often noisy flight and, on alighting, scramble or hop among the branches reaching out for fruit, small animals, or both. The wood hoopoes have weak flight, and they do not fly much; they fly chiefly from one tree or clump of trees to the next, climbing about the trunks and branches of trees and lianas in acrobatic poses as they seek insects in crevices and on the bark surface.

Walking on the ground is the usual mode of feeding for the common hoopoe (*Upupa*), the ground rollers, and for a few hornbills. The hoopoe walks with quick steps, bobbing its head in time with the steps and pausing to probe with its long bill in the ground and in crevices, in search of large arthropods and small vertebrates. Its flight is strong and direct. When perched, it may quietly flash its long crest open and shut. The ground rollers, most of which are birds of the deep forest, also feed on the ground on food similar to that of the hoopoe. When disturbed, they fly or jump to low perches.

Social feeding by ground hornbills

The ground hornbills (*Bucorvus* species) exhibit a definite social organization when foraging. Three or four members of a group searching for insects and other small animals on the ground may keep near each other, with the result that prey frightened into activity by one bird may be caught by one of the others. Several other species of hornbills occasionally forage solitarily on the ground.

Habitat selection and ecological diversity. Coraciiform birds, diverse in their structure and behaviour, occupy a variety of habitats. Each species is restricted in distribution by requirements of feeding and nesting areas. A species may be considered to occupy a feeding and, during the breeding season, a nesting habitat, and these may or may not be contiguous. Some species, such as certain African kingfishers, which nest in cavities excavated in termite mounds and feed on termites and other insects near the nest site, may be said to nest within their feeding habitat. Aerial feeders, such as bee-eaters, which nest in burrows, may be considered to occupy two habitats, one for feeding and one for nesting.

The factors that influence habitat suitability are evident in a few cases. One bee-eater requires a cut bank in a grassland area for its burrow, but must be near a forest because it feeds over the forest and the forest edge. The importance of suitable perches is illustrated by observations made on an insect-rich region of East African grassland, from which kingfishers were absent until a road was built. With the roads came telegraph wires, which provided the perches.

Relative diversity of coraciiform families

The widest range of habitats occupied by members of the order Coraciiformes is found in Africa (including Madagascar), where eight of the 10 families are represented. These exhibit all four basic feeding modes, but on Madagascar some niches are occupied by families different from those on the African mainland. Ground feeding in the forest, for example, is limited to the ground rollers in Madagascar; in Africa this niche is occupied by the hoopoe (except in dense forest) and by birds of other orders. The four coraciiform families found in temperate Eurasia occupy only three niches: ground feeding (hoopoe), darting from a perch (kingfishers, some bee-eaters), and aerial hawking (other bee-eaters, rollers). In Australia only the last two niches are occupied by coraciiforms, and these are the same families that hold these niches in Eurasia. In the New World, where three families are found, only one feeding mode is used, that of darting out from a perch; but this mode is geographically and ecologically subdivided between the diminutive todies, which are limited to the Greater Antilles where neither the kingfishers nor the motmots are represented, and the kingfishers (with the exception of the pygmy kingfisher, *Chloroceryle aenea*), which are specialized fish eaters.

Social behaviour. A number of coraciiform birds are markedly social, feeding in small parties and nesting in colonies. Some wood hoopoes forage in conspicuous, noisy bands of five to 10 individuals. The acrobatic, climbing activity of a band is sometimes interrupted when the birds of a whole party bow and sway their bodies, pump their tails up and down, and join in a chorus of chattering calls. Moving from tree to tree, one bird follows another in weak undulating flight. Forest species of wood hoopoes are less social, and lone individuals sometimes call while perched high in a tree. Many hornbills may be seen flying through or over the forest, the beats of their broad wings giving a characteristic loud whooshing noise. They are usually found in small parties and actively move about in the branches, sometimes giving conversational notes. A study of a young captive hand-reared hornbill of a small species of *Lophocerus* (*Tockus*) has provided surprising data. This bird seemed to have a remarkably active, alert, intelligent personality recalling that of captive crows. It greeted its foster parents by raising its wings, pointing its beak upward, raising its head feathers, and chattering. It was jealous of attention given other animals, kept close to its foster parents when out of doors, and alighted on their shoulders. When accompanying them on walks, it flew from tree to tree. It was busy and mischievous, attracted by anything bright, and fond of picking at knots or holes. This young hornbill had a passion for pulling up seedlings and sometimes amused itself by darting in and tweaking the tail of a larger, more lethargic, young *Bycanistes* hornbill.

Relationships with other species. Certain types of social behaviour of the Coraciiformes involve other birds or unrelated animals. Although some of these interactions are occasional and opportunistic, others are regular parts of everyday life and may be called symbiotic—*i.e.,* one that brings mutual benefit to the different species involved.

Utilization of grass fires

The regular swarming of many bird species about grass fires to capture animals driven out of hiding by the flames is a phenomenon often related to human activity, for such events are often caused accidentally or deliberately by man. Among the birds that gather are both rollers and bee-eaters; they swoop down near the flames and into the smoke to seize fleeing insects. After the fire has passed, certain hornbills find good foraging on foot over the newly exposed ground.

More notable are a number of interspecific nesting relationships. Some bee-eaters make their colonial burrows in the same banks in which certain smaller swallows dig their burrows; there seems to be no conflict between the larger bee-eaters and the smaller swallows, despite the similarity in nesting and feeding habits. In southern Africa, the little bee-eater (*Melittophagus pusillus*) sometimes makes its nest burrow in the wall of the very much larger burrow of the aardvark (*Orycteropus afer*), and there is no further relationship between the bird and the mammal.

Sporadic incidents occur between species when one or both are foraging; a kingfisher may pilfer a food item from a dipper (*Cinclus*), and a savanna kingfisher will occasionally fly down to seize a grasshopper flushed by a man. Many associations are more frequent. Some bee-eaters in Africa often accompany large bustards, other large walking birds, and zebras and other game animals to feed on the insects roused from the grass by the animals. The bee-eater even uses the bustard's back as a perch. The bee-eater may also accompany an automobile driven through these grasslands to secure insects. There is also a regular association between hornbills and bands of monkeys in the treetops of African forests, with the birds seeking the insects stirred into activity by the fruit-eating monkeys.

In many parts of the Old World tropics, where large arboreal termite mounds are common and conspicuous, certain kingfishers usually excavate their burrows in them; in fact, some species are believed to nest only in them. The presence or absence of the termites might be expected to have an important effect on the populations of such kingfishers. The hoopoe commonly nests near buildings, especially in South Africa, and it is possible that the availability of such sites may affect the local abundance of the species. The presence of woodpecker holes used by hoopoes and wood hoopoes may also affect the size of their breeding populations.

Some African species of kingfishers, bee-eaters, hoopoes, and wood hoopoes are victimized by obligate social parasites, the honey guides (Indicatoridae, related to the woodpeckers). The honey guide lays its eggs in the host's nest and, with its bill or claws, often punctures the shell of the foster parents' eggs so they do not hatch. If the foster parents' eggs do hatch, the nestling honey guide usually disposes of the host's young by throwing them from the nest or by biting, crowding, or starving them to death. The honey guide's young are thus raised at the expense of the young of the host species. Apparently kingfishers, bee-eaters, hoopoes, and wood hoopoes are of great importance in the ecology of honey guides, and the frequency with which the roller-like birds are victimized by honey guides may be a serious factor in their population status.

A remarkable insect fauna has been found in the nest of an African hornbill. Though some nest sanitation is practiced by the birds, it is not complete. In one nest, more than 400 individual insects, mostly larvae, were found (about half were moth larvae); they represented eight species and were feeding on the droppings and debris in the nest cavity, which was remarkably clean and had little odour. The hornbill provides microhabitats for the insects (albeit scattered and seasonal), and the scavenging of the insects may be of advantage to the hornbill.

FORM AND FUNCTION

Size and plumage. The coraciiform birds are a rather heterogeneous order, united mainly by features of their internal anatomy. Some characteristics of the beak and feet serve to separate them from other orders, such as perching birds (Passeriformes) and the woodpeckers and their allies (Piciformes), which appear to be their closest relatives.

No single coraciiform family encompasses the entire size range of the order. The todies are the smallest, with lengths of nine to about 11.5 centimetres (3.5 to 4.5 inches), and the hornbills, from about 40 to 160 centimetres (16 to 63 inches), are the largest. The kingfishers are from 10 to nearly 46 centimetres long (four to 18 inches), the longest being those with extended tail feathers. Motmots and bee-eaters are in the same general size range as the kingfishers, but the smallest of them are larger than the tiniest kingfishers (*Ceyx, Ispidina*), and the largest motmots, although about 50 centimetres (20 inches) long, have not nearly the body bulk of the chunkier, but slightly smaller, kookaburras (about 45 centimetres, or 18 inches). The smaller families have, predictably, less size variation.

Ornamental coloration

The plumage of the roller-like birds is firm and often highly colourful. The bee-eaters are collectively and individually among the most brilliantly coloured of all birds; one individual may be marked with green, yellow, red, blue, and black. Many kingfishers are also brightly coloured, with a tendency toward metallic blues and blue greens. The beak is often bright red or orange. Most hornbills, with ornamentation frequently found on the beak, are strikingly patterned in black, white, and shades of gray and are sometimes accented with rufous or yellow; many have areas of bare skin, blue, red, yellow, or black in colour, around the face.

Morphological specializations. A few of the external features, such as modifications of the wings and feet for locomotion and of the beak for food handling, are obviously related to behaviour and habitat. Even in these important aspects of the body plan, the common heritage is evident in such features as the small feet and fusion of the front toes.

The most obvious adaptation to behaviour is the shape of the wing. The size and shape of the wing correlates well with the type of flight. Aerial feeders have the longest, most pointed wings; the most extreme forms are found in the bee-eaters, but they are also well developed in the rollers and the cuckoo roller. The birds that watch for their prey and fly out after it (*e.g.,* kingfishers) have moderate, rounded wings, while the ground feeders (ground rollers, hoopoes) have broad wings, as do those that feed on foot in the trees (hornbills, wood hoopoes).

The tail is highly diversified in length and shape. Forked tails occur only in the best fliers (bee-eaters and rollers), though some of these birds have square tails or elongated central tail feathers instead. Elongated central tail feathers also occur in the hornbills that practice direct flight and in ground rollers that fly little. The small kingfishers have the shortest tails. The exact shape of the tail does not correlate well with locomotion, nor does the presence of spatulate tips on elongated tail feathers in motmots, rollers, and kingfishers, which are perhaps of social importance.

The characteristic short tarsus (the lower leg) of the order and the fusion of the three front toes (except in the cuckoo roller) seem a heritage that has been modified but little with modification of behaviour. The foot is used only for perching in most family groups. In cases in which it is used extensively for terrestrial locomotion (ground rollers, hoopoe), the tarsus is lengthened somewhat in the former but not in the latter, and is also somewhat lengthened in the aberrant hornbill, known as the ground hornbill. The hornbills that feed in the tree branches have a broad pad on the toes, evidently an adaptation for perching. The wood hoopoes have long slender toes with long, sharp, curved claws, an obvious adaptation for the bark-climbing habits of these birds.

Drawing by R. Keane

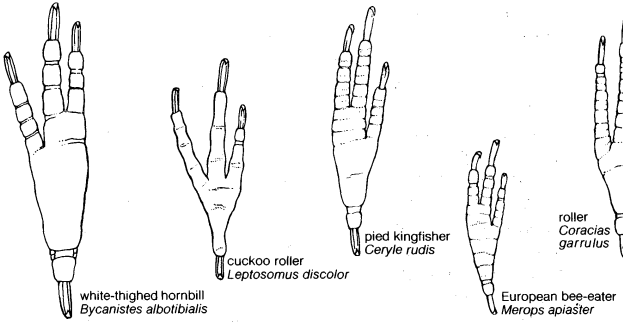


Figure 43: Feet of some Coraciiformes, plantar (bottom) view.

The most unusual foot in the order belongs to the cuckoo roller, whose outer toe is capable of being reversed. This makes a better perching foot for the bird that flies over the forest trees, scanning the branches for its prey, and alighting suddenly to seize a caterpillar, chameleon, or grasshopper.

Diversity in the bill

The bills show remarkable diversity in bulk and shape. Basically, the long stout bill so common in this order seems to be an adaptation for seizing and subduing active animal prey that is large in proportion to the size of the bird. This is true for rollers, ground rollers, cuckoo rollers, and motmots, all of which have only moderately long and stout bills. The larger, rather stouter, straight bill of the kingfishers is an exaggerated version in birds that often take small invertebrates. The hornbills, with their very large, laterally compressed bills, often somewhat ornamented with a prominent horny casque in the male (smaller in the female), seem to have carried bill size beyond the point of a strictly functional feeding organ. Species that feed largely on fruit plucked from branches as well as species that take lizards and snakes and dig in the ground for insects all have exaggerated bills.

An advantage of a long bill can be seen in hornbills that feed on fruits among the outer branches of forest trees; the long bill enables the bird to reach fruit on slender, outer twigs. There is also the probability, evidenced by the sexual difference in bill ornamentation, that the bill serves in courtship and perhaps in other social contacts within the small parties characteristic of hornbills.

Another type of long bill in the order is that of the hoopoes and the wood hoopoes, which is a slender and slightly, to strongly, downcurved bill. The former use the bill to probe in the ground while walking, the latter to poke into crevices and crannies of the barks and branches of trees.

Similarities to birds of other orders. The size of the birds of this order and their propensity to take rather large prey bring them into competition with many other species of other groups. It is perhaps instructive to compare certain Old World groups with representatives of different orders in the New World tropics (see above *Natural history*). The most striking parallel is seen between the toucans (Rhamphastidae, order Piciformes) and the hornbills, which, with their enormous bills, small feet, general diet, behaviour, and appearance, are remarkably alike. A similar degree of convergence is seen between the bee-eaters and the jacamars (Galbulidae, order Piciformes) of tropical America and also between some African wood hoopoes (*Rhinopomastes*) and American wood hewers (*Campylorhamphus*). Similar ecological conditions apparently have brought similar adaptations in external structure and behaviour between birds of quite unrelated orders, in widely separated areas.

EVOLUTION AND CLASSIFICATION

Evolution. The history and some of the family relationships within the Coraciiformes are obscure, so that any treatment of the evolution of the order must be considered speculative. The present distribution and abundance of the 10 families suggest an Old World origin, probably in the Ethiopian–Indian region, but the southern Palaearctic (Eurasia) may also have been involved; the limited number of fossils, consisting of a few hornbills from the Eocene (about 50,000,000 years ago) and Miocene (about 15,000,000 years ago), a roller from the upper Eocene or lower Oligocene (about 38,000,000 years ago), a possible wood hoopoe from the Miocene, and a kingfisher from the lower Oligocene, all have migrated from Europe. At the periphery of the Ethiopian region, the island of Madagascar has the endemic ground-roller and cuckoo-roller families, probably derived from separate colonizations of the early roller stock; the island was later colonized by the modern rollers, bee-eaters, kingfishers, and the hoopoe. Extending eastward through southern Eurasia, the modern hoopoe (*Upupa epops*) reaches Malaysia (making it the most widespread single coraciiform species); the hornbills have reached the Papuan area; and the roller, bee-eater, and kingfisher families have reached the Australian continent.

Kingfishes advancement in the New World

In the New World, the early arrival of a protokingfisher stock, via the Bering Strait, probably gave rise to the motmots of Central and South America and the todies of the West Indies. Later, a specialized, fish-eating branch of the kingfishers colonized the New World, evolving one species in the Nearctic region (North America) and several in the Neotropical region.

Considering the relative paucity of roller-like birds in tropical America and their comparative abundance and diversity in the Old World tropics, it seems likely that, by the time the coraciiform stock had reached the neotropics, many niches occupied by members of this order in the Old World were already filled by members of various piciform and passeriform families. The passeriform suborder Tyranni, with more than 600 New World species, is particularly diverse. The presence of highly adapted potential competitors, such as the toucans, jacamars, and puffbirds, endemic members of the Neotropical avifauna, may have retarded the colonization and evolution of the Coraciiformes in the New World.

Classification. *Distinguishing taxonomic features.* The external characteristics on which families are based are the size and shape of the beak and wing and the arrangement and amount of fusion (syndactyly) of the three front toes. The ten families are united, additionally, by features of the palate bones, the tendons of the leg, the configuration of the leg muscles, and the body pterylosis (pattern of feathers).

Annotated classification. The following classification is based on an arrangement used in 1960 by Alexander Wetmore, a U.S. ornithologist.

ORDER CORACIIFORMES (roller-like birds)

Small to medium-large land birds, with body lengths of about 10 to 160 cm. Most species conspicuously coloured. Beak prominent; straight, or slightly or strongly downcurved. Some syndactyly of toes (I, II, and III) in most families. Cavity nesters; young hatched blind and naked (except in Upupidae). Worldwide in temperate and tropical regions. About 191 species.

Family Alcedinidae (kingfishers)

Oligocene to Recent. Chiefly arboreal; short tarsus, small feet; syndactyl. Beak medium to long, straight, stout, usually spearlike. Wings short, rounded. Food: invertebrates and small vertebrates, including fish. About 90 species; distribution that of order, but greatest diversity in Indo-Australian Region; length 10–45 cm.

Family Todidae (todies)

Recent. Chiefly arboreal. Long, straight, flattened, blunt bill. Toes syndactyl. Wings short, rounded. Food: invertebrates, insects. Five species; West Indies; length 9–12 cm.

Family Momotidae (motmots)

Eocene to Recent. Chiefly arboreal. Moderately long, stout, pointed, slightly decurved bill usually with serrate cutting margins. Tarsus very short, toes syndactyl. Wing short and rounded. Food: invertebrates, lizards, and some fruit. About 8 species; South and Central America; length 17–50 cm.

Family Meropidae (bee-eaters)

Pleistocene to Recent. Arboreal and aerial. Bill long, compressed, tapering to a fine point, and slightly decurved. Tarsus short, anterior 3 toes slender, weak, and syndactyl. Wing long and pointed. Food: insects. About 24 species; Africa, southern Eurasia to Australia; 15–35 cm long, including elongated tail feathers.

Family Coraciidae (rollers)

Eocene to Recent. Chiefly arboreal and aerial. Bill stout, crowlike, slightly downcurved, terminally hooked. Tarsus short, foot strong; inner and central toes united at base. Wing long, moderately pointed. Food: chiefly insects. About 12 species; temperate and tropical parts of the Old World, greatest number of species in Africa; length 25–32 cm.

Family Brachypteraciidae (ground rollers)

Recent. Chiefly terrestrial in forest and desert brush. Roller-like birds with longer tarsus and short, rounded wings. Food: small animals of forest floor or desert brush. Five species; Madagascar; length 30–40 cm (including long, graduated tail of some species).

Family Leptosomatidae (cuckoo roller)

Recent. Arboreal and aerial. Bill moderately long, stout, slightly decurved, and terminally hooked. Tarsus very short and, unique in this order, toes semi-zygodactyl (the outer, anterior toe reversible). Wings long, moderately broad and somewhat pointed. Also unique in having a pair of powder-down patches, one each side of rump. Food: large insects, lizards. One species; Madagascar; length about 43 cm.

Family Upupidae (hoopoe)

Pleistocene and Recent; terrestrial and arboreal. Bill long, slender, slightly decurved. Tarsus short, slender; toes, long, with central and outer ones fused at base, claws short. Wing moderate, broad. Food: arthropods, and other invertebrates. One species; Africa, southern Eurasia, and Malaysia; length about 29 cm.

Family Phoeniculidae (wood hoopoes)

Miocene (Europe), Recent (Africa); arboreal. Bill long, slender, slightly curved to sickle-shaped. Tarsus very short; toes long, central and outer ones fused at base. Claws long, curved and sharp. Food: invertebrates chiefly. Six species. Length 22–38 cm.

Family Bucerotidae (hornbills)

Eocene (Europe) to present. Chiefly arboreal (1 species chiefly terrestrial). Large, slightly curved bill, often with casque or sculpturing (larger in males). Tarsus short to very short, toes syndactyl. Wings moderate to long and broad. Unique in order in having eyelashes. Food: insects, small vertebrates, and fruit. About 45 species; Africa, southern Asia to Papuan area; length 40–160 cm.

Critical appraisal. Paleontologists have called the order Coraciiformes a miscellaneous assemblage of birds with

Divergence among families

large bills and have pointed out that the anatomical differences between the families of birds within this and other orders are minor compared with differences found within some mammalian orders, such as the carnivores, which include both cats and seals. The classification of birds into higher categories is justified by expediency and tradition. The roller-like birds include ten groups of birds here called families, the largest, the kingfishers and the hornbills, containing 90 and 45 species respectively, the smallest, the hoopoe and the cuckoo roller, a single species each. Each family is fairly uniform within itself, but between families some anatomical characters have a disconcerting habit of lacking a common linkage. The order does seem to have a common heritage, however, judging from the sum of the characters that place them high in degree of specialization, just below the perching birds (Passeriformes) and the woodpecker-like birds (Piciformes). They have been variously said to be distantly related to the cuckoos, parrots, owls, and even the nightjars, depending on the relative weight given to characters such as the condition of the bony palate, the notching of the sternum, the arrangement of the toes, the thigh muscles and foot tendons, the structure and location of the syrinx (sound-producing organ), the development of the ceca (blind pockets in the digestive tract), and the arrangement of the intestines, the arrangement and type of feathers and down, and the type of nesting.

Some authorities would include rollers, ground rollers, and cuckoo rollers as subfamilies of the Coraciidae. There is also a question as to whether the hoopoes and wood hoopoes are more nearly related to the hornbills or the rollers.

The families fall into six or seven well-defined groups: (1) kingfishers, todies, and motmots, (2) bee-eaters, (3) rollers and ground rollers, (4) cuckoo rollers, (5) hoopoes, (6) wood hoopoes, which are sometimes united with the Upupidae, and (7) hornbills. (Au.L.R.)

Piciformes (woodpeckers, barbets, honey guides, toucans)

The bird order Piciformes includes the familiar woodpeckers, their relatives the piculets and wrynecks (collectively, the family Picidae), and the exotic tropical jacamars (Galbulidae), puffbirds (Bucconidae), barbets (Capitonidae), honey guides (Indicatoridae), and toucans (Ramphastidae). This arboreal group of about 370 species is distributed on all continents except Australia and Antarctica, but only the woodpecker family is widespread outside the tropics. Although the six families comprise the order, over half of the species represent one subfamily (Picinae, true woodpeckers) of the family Picidae. The order includes such familiar birds as the European great spotted woodpecker (*Dendrocopos major*) and green woodpecker (*Picus viridis*) and the American flicker (*Colaptes auratus*) and downy woodpecker (*Dendrocopos pubescens*). Piciforms are economically important because they include many insect-eating species. A few species are harmful in eating fruit (toucans) or damaging trees (sapsuckers, genus *Sphyrapicus*), but even these species eat some insects and other animals, and they of course serve as natural balances within the ecological systems of which they are a part.

Distribution of the order

The tropical distribution of most piciform families suggests that these represent specialized remnants of a once more numerous and diverse array of arboreal birds. Perhaps many elements of the order became extinct during the geologically recent burgeoning of the diverse, ubiquitous, and highly successful order Passeriformes, which the piciforms resemble in many ways. It has even been suggested that both groups be merged into one order.

Piciform species vary in size from about nine to more than 60 centimetres (3½ to 24 inches) in overall length. They vary greatly in the structure of their beaks and only slightly less in the rest of their morphology, some, such as the huge-billed toucans and the sturdy woodpeckers, being very specialized. The specialized habits of the wax-eating honey guides are unique among birds. The most numerous and widely distributed groups, the barbets (70 species) and the woodpeckers (about 203 species), excavate their own nesting cavities, thereby avoiding competition with other birds and, incidentally, providing homes for many other species of vertebrates.

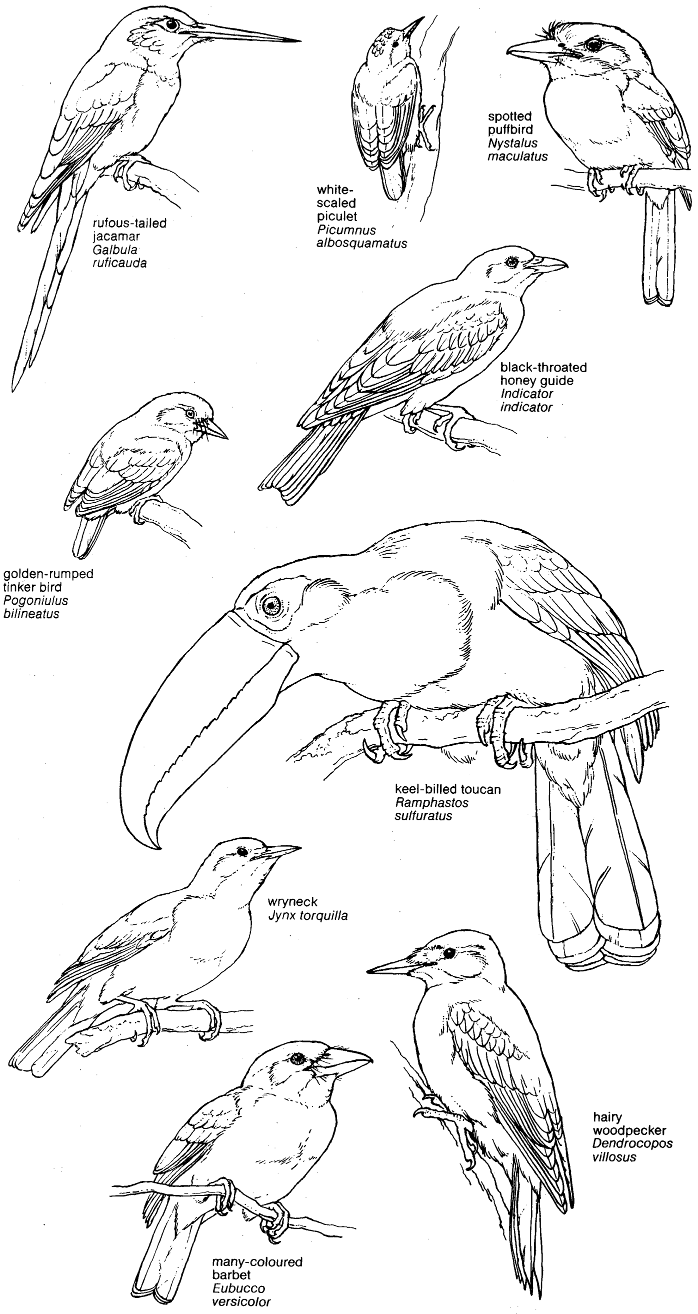


Figure 44: Body plan of representative piciform birds.
Drawing by R. Keane

IMPORTANCE TO MAN

Most piciform birds consume insects, some foraging in places (within bark) reached by no other birds. Thus, they are valuable in the control of insects, even helping to prevent the spread of tree diseases, such as Dutch elm disease, by destroying insect carriers. Fruit-eating barbets and toucans spread seeds of certain tropical trees, the seeds being passed through their bodies at distances away from the parent tree. Honey guides lead man, as well as other mammals, to sources of honey. Colourful toucans, barbets, and jacamars, and even the dull-coloured puffbirds delight both natives of and visitors to the tropics. Woodpeckers are conspicuous and often colourful species of the forest, and their actions have long elicited interest and admiration from man—their winter visits to suet feeders in the Northern Hemisphere are anticipated and enjoyed by millions of onlookers. The excavation of nesting and roosting holes by woodpeckers and barbets provides homes for countless birds of other species, which utilize the holes after the excavators are finished with them. Foresters in Europe frequently compensate for the lack of old woodpecker holes needed by valued insect-eating species by placing artificial nest boxes in their cultivated forests.

Artificial nest boxes

Fruit eating by barbets and toucans sometimes causes damage to cultivated fruit plants, especially near villages and towns. The few species of sapsuckers in North America and southern Asia occasionally destroy trees by creating openings through which disease agents may enter, although wounds in healthy trees rapidly heal. Piciform birds that excavate their nests usually choose dead or dying trees, but some excavate in live trees or cacti. The area around the excavation may be sealed off by scar tissue in some trees, especially cacti, resulting in no damage but diminishing the tree's value. Some woodpeckers damage fence posts and utility poles, particularly in open country, in well-kept woodlots, and near cultivated forests where dead trees are scarce.

NATURAL HISTORY

Distribution and ecology. Piciform birds are varied in their habits, although certain facets of behaviour relating to hole nesting, such as the snakelike hissing of young birds in the nest, are shared with some other hole-nesting birds. Jacamars and puffbirds, the fly-catching members of the order, are stolid, unsuspicious birds, allowing close approach by humans. Woodpeckers, toucans, and barbets are more wary, but the smaller species of piculets and woodpeckers sometimes can be approached closely.

Jacamars and puffbirds inhabit various portions of tropical American forests. Jacamars sit quietly on perches in openings and edges, flying infrequently, but often flicking their tails. Large insects, such as butterflies, are captured, taken to a perch, and usually beaten against it. The wings of the insects are removed if the prey are to be fed to the young.

Puffbirds are often found in farm areas, where they may be seen along the road looking something like dull-coloured forest kingfishers. They have a rapid, darting flight, and some species, such as the swallow-winged puffbird (*Chelidoptera tenebrosa*), spend much time making sorties after insects high in the trees. Others forage low in bushes, obtaining prey on the ground as well as in undergrowth. Tail-jerking movements are observed in some puffbirds (*e.g.,* the two-banded puffbird, *Hypnelus bicinctus*), as in jacamars.

The barbets are weak-flying, essentially resident birds inhabiting woodland and savannas, occasionally the fringes of deserts. They feed on insects, fruit, and even flowers and buds. The tiny African tinkerbirds (*Pogoniulus*), so named for their repetitive vocalizations, feed mainly or entirely on insects, foraging rapidly over bark and occasionally tapping with the beak, very much like piculets of the woodpecker family. Some barbets have serrated, or "toothed," edges of the bill, in the manner of toucans; these are apparently used for tearing open fruit.

Guiding by honey guides

Honey guides are nonmigratory birds, dwelling in pairs or alone in forests and savanna. They feed on insects, fruit, and beeswax. Only a few honey guides of the genus *Indicator* actually are known to "guide" certain mammals, including man, to bees' nests. Most honey guides are able to enter bees' nests without assistance, there to eat larvae and, primarily, wax, which is digested with the help of special bacteria within the birds' digestive tracts. Honey guides that "guide" use a peculiar chattering to attract the attention of an animal likely to follow. The greater honey guide (*Indicator indicator*) spreads its tail conspicuously and uses a peculiar undulating flight in attracting attention. It perches and calls again, then moves further on in another short flight, as the interested animal draws near. Upon reaching the bees' nest, the honey guide waits quietly in the vicinity until the animal has taken the honey and moved off. It then flies to the devastated nest and feeds on scattered pieces of wax and exposed larvae.

Toucans are strictly New World tropical-forest birds, although they may forage about farms. One group, the hill toucans of the genus *Andigena,* occurs in the Andes up into temperate forests. No toucans are geographically migratory, but some species make local movements up and down mountain slopes. Toucans eat much fruit, perhaps aided by their serrated beaks. Smaller pieces or small fruits are first held in the tip of the bill and then thrown back into the mouth. They also eat larger insects, birds' eggs, and even young birds. The colour patterns of the bill are distinct in each species and probably serve a behavioral function. Toucans have a peculiar habit of cocking the tail forward over the head, particularly while sleeping in a cavity; D'Arnaud's barbet (*Trachyphonus darnaudii*) also does this during displays.

The family Picidae is comprised of three distinct subfamilies. One, the Jynginae, contains two closely related species of wrynecks (*Jynx*), found in Eurasia and central Africa. Wrynecks, which have soft, unspecialized tails, forage for ants in trees and on the ground.

The piculets (Picumninae) comprise a second subfamily of soft-tailed, woodpecker-like birds. All but four of about 30 species live in forests and dry woodlands of the New World tropics. The major genus *Picumnus* is peculiarly distributed in having about 24 species in the New World and one species (*Picumnus innominatus,* the speckled piculet) in southern Asia, a pattern of distribution matched by the woodpecker genus *Celeus.* All the piculets are tiny, the largest—the peculiar Haitian piculet (*Nesoctites micromegas*)—being only sparrow sized. They feed in a woodpecker-like manner, tapping and probing usually on small branchlets and vines, and they allow close approach, continuing to forage under the eyes of an observer. Their food is composed mainly or entirely of insects.

Feeding behaviour of woodpeckers

The true woodpeckers (Picinae), with their stiffened tails, usually perch lengthwise on tree branches or trunks. Among the relatively few species that go to the ground, all hop, except for two species of flickers (campo flicker, *Colaptes campestris,* and Andean flicker, *Colaptes rupicola*) that walk. Foraging is by diverse modes. True excavating, which penetrates the substrate of the bark, is not as common as one would expect on viewing the woodpecker bill. Most woodpeckers forage mainly by gleaning items from the surface of a tree, by probing into cracks, and by occasional to frequent tapping, sufficient to break barely through the surface. Only the heavy-billed, so-called ivorybills (genus *Campephilus*) of the New World, some of their relatives, other similarly specialized Asian groups, and a few of the smaller woodpeckers (genus *Picoides;* some species of *Dendrocopos*) regularly chisel deep into trees. The food of woodpeckers is varied, mainly consisting of insects, but some species (especially of the genus *Melanerpes*) frequently eat fruits and berries. Ants are a favourite food, especially in tropical regions—fully half of the numerous tropical species feed largely or entirely on ants. Three woodpeckers (ground woodpecker, *Geocolaptes olivaceus* of Africa, the campo flicker, and the Andean flicker) are ground dwellers, feeding mostly on ants, while seven or eight other species frequently forage on the ground. The flight of woodpeckers is strong, with definite up–down undulations.

Most woodpeckers are permanent residents, the few exceptions including the North American yellow-bellied sapsucker (*Sphyrapicus varius*) and flicker, which are highly migratory.

Reproduction. The ability of most piciform birds to excavate their own nest holes is probably a major factor responsible for their success. With relatively diverse types of bills, the jacamars, puffbirds, and barbets excavate nesting cavities, as do the sturdier billed woodpeckers, although the former utilize a softer substrate such as earth or dead trees. Woodpeckers alone within the order have biologically "capitalized" on their adaptations for excavating nests, by various further specializations for nesting (even in hard, living trees) and for foraging on the surface and within the bark of trees. They have thus exploited niches unavailable to other birds, a fact that accounts for their success relative to other members of the order. Yet because of their specialized habits, only a limited number of kinds of woodpeckers can exist together, usually no more than five to eight species (more in some tropical environments), differing somewhat in size and habits.

Difficulties in the study of hole-nesting birds

Because of the concealing nature of the nest cavity, the behaviour of hole-nesting birds is more difficult to study than that of species building exposed nests. The nesting behaviour of most piciform families must be inferred from observations on a few examples, the woodpeckers being the only group that has been extensively studied.

Jacamars are usually resident wherever they occur and nest at diverse times of the year depending on the locality. Nests are excavated in banks or in steep hills. Both sexes dig with the beak, kicking out dirt with their feet. The tunnel is comparatively short, usually less than one metre (three feet) in depth. No lining is placed in the nest chamber at the end of the tunnel. Three or four white eggs are incubated by both sexes, with the female incubating at night. The incubation period of several species of jacamars is known to be from 19 to 21 days. The young are downy and hatch with their eyes closed. They remain for 20 to 26 days in their nest before leaving. Most of the food consists of large insects brought to the young by both parents. The parents do not clean debris and fecal material of the young from the nest, which becomes littered.

Puffbirds excavate their nests in the ground or in termite nests on the ground or in trees. The tunnel is usually deeper than that of a jacamar, often reaching a metre or more into the ground. Several pairs may nest close together, semicolonially. Both sexes excavate the burrow. Some species leave the chamber at the end of the tunnel unlined, while others may line it with grass or dead leaves. Several species, including the white-whiskered puffbird (*Malacoptila panamensis*), pile sticks and leaves at the mouth of the nest tunnel, effectively concealing it. Both sexes incubate the two (occasionally three) white eggs. The newly hatched young are naked and are fed by both parents.

Social nesting and roosting

Although the majority of barbets nest in isolated pairs, there is a marked trend toward sociality. Large numbers of one or several species may gather to feed at a fruiting tree. Some species are found in family groups, and displaying groups of several individuals may be seen during the onset of breeding. The most social of all are the plain-coloured barbets (genus *Gymnobucco*) of Africa, the brown barbet (*Calorhamphus fuliginosus*) of southeastern Asia, and some species of the genus *Megalaima* of southern Asia. These nest colonially and feed in gatherings of up to 50 or more birds (*Gymnobucco*). They are aggressive, noisy birds and may take over an entire dead tree for nesting, driving away other bird species. In some barbets of the genera *Semnornis* and *Megalaima*, communal roosting occurs, with numbers of birds spending the night together in the same hole. Nests are usually excavated in trees, but a few species burrow into the ground or into termite nests. Both sexes excavate the cavity, which typically has a nearly round entrance, small for the size of the bird. The same cavity may be used for several nesting seasons. No lining material is placed in the nest cavity. The shape of the barbet's bill, which is not at all chisel-tipped like that of a woodpecker, makes the excavation of cavities in trees a task requiring great effort, and it appears that only dead, well-rotted trees are used. Two to five white eggs are laid in the cavity; both parents then incubate the eggs for 13 to 15 days or more. Both also may spend the night in the nest. The young develop slowly, spending 30 or more days in the nest. The adults clean the nest, removing excreta. Often the young and adults continue to roost together after the young have left the nest, occupying a separate "dormitory" hole excavated for that purpose. On occasions, more than one pair of birds may use the same nest (white-eared tinkerbird, *Pogoniulus leucotus*).

Most, if not all, of the honey guides are brood parasites, laying their eggs in the nests of other species, especially those that nest in holes or in deep, covered nests. The host's eggs often are punctured by the female honey guide when she lays her own egg. Only a single egg is laid in each host's nest. Because of the parasitic habit, the incubation periods of honey guides are not well known, but periods of 12 to 16 days have been found in species of *Indicator*. The young birds spend a long time in the nest, 35 to 40 days or more. When newly hatched, they have hooks on the tips of their bills, which are used to injure and kill the young of the host species. These hooks later fall off.

Toucans nest in natural cavities or in old woodpecker nests. Smaller species may usurp a nest freshly constructed by woodpeckers. Old, well-rotted cavities may be enlarged slightly. One to four eggs are laid on the floor of the nest, which may be unlined or lined with a few leaves. About 16 days are required for incubation, performed by both sexes. The birds are shy about the nest and hence difficult to observe. The young, naked when hatched, bear large pads on their heels, on which they perch in the nest. The nestling period is generally long, up to seven weeks or more, varying with the size of the birds. Toucans, except the toucanets (genus *Aulacorhynchus*), do not maintain clean nests. Family groups may remain together for a long time, for small flocks are often seen throughout the year.

Wrynecks usually nest in natural crevices, old woodpecker holes, or old swallow holes in banks. They lay two to 14 eggs, and both sexes incubate. Like the American flicker and probably other woodpeckers, the wryneck is an indeterminate layer; *i.e.,* the bird produces eggs until a certain number have accumulated.

The nests of piculets are constructed in small trees, stubs, fence posts, and the like. The entrance is tiny and rounded, resembling a barbet nest entrance. Two to four eggs are laid and incubated for about two weeks. The initially blind and naked young grow for three weeks or more before leaving the nest for the first time. Both parents incubate and feed the young, and in at least some species they both remain in the nest at night. After fledging, the young may continue to roost in the nest at night with their parents for several months. Like the wrynecks and most other birds and unlike woodpeckers, the piculets often perch crosswise on branches.

The reproductive behaviour of woodpeckers is better known than that of most other Piciformes. The sexes are clearly marked by simple pattern differences, most commonly involving the presence or absence of red on the head. Observations at bird-feeding stations suggest that sexual recognition occurs throughout the year, with males dominant over females. Aggressive displays are conspicuous and numerous in woodpeckers, including various head postures centring on the bill, bill pointing, head turning, swinging of the head and body from side to side, bobbing and bowing of the head, wing flicking and spreading, tail spreading and turning, and even aerial displays. These displays exhibit the complex head markings, as well as the body, tail, and wing patterns found in most species. Many vocal and instrumental signals are used, including the widely prevalent tapping or drumming of the bill against a tree or, in cities, on tin roofing.

Displays of woodpeckers

Most woodpeckers actively defend territories, often throughout the year. This perhaps accounts for their strongly developed aggressive behaviour. Few species are markedly social, but members of the New World genus *Melanerpes* especially tend to be, even while nesting. Up to 11 different adults of the yellow-tufted woodpecker (*Melanerpes cruentatus*) were observed feeding young in three different nests in eastern Peru. Some of the adults fed young in two and even in all three nests. Andean flickers nest in loose colonies in banks and may be seen in groups of 10 or 12 birds. Otherwise, social gatherings of woodpeckers generally involve family groups during and sometimes after the breeding season. Most woodpeckers are aggressive toward others of their own species, even to some extent toward their mates and young. Aggressive encounters are most frequent between members of the same sex. When an intruding woodpecker invades the territory of a pair during the breeding season, it usually is attacked by the member of the pair that is of the same sex as the intruder. Conflicts between birds after the breeding period are more intense when the woodpeckers involved are of the same sex. Experiments with the North American flicker show that even paired birds recognize each other primarily by their sexual colour markings, for a female given an artificial black moustache stripe (characteristic of the male) was attacked and pursued unceasingly by her mate until the experimenter rescued the exhausted female and removed the "moustache." She was once again fully accepted by her mate.

Nests are excavated by the birds themselves, although a

Nesting of woodpeckers

few species will accept artificial nest boxes. Two species, the ground woodpecker and the Andean flicker, always nest in the ground, digging nests in banks, and the campo flicker of South America often does so. Other woodpeckers chisel holes in trees, fence poles, cacti, or similar sites. Usually both sexes construct the nest, but either sex may do most of the work, depending on the species. The nest cavity varies with the size of the bird. A flicker (30 centimetres, or 12 inches, long) was found incubating eggs in a tree stub only 12 centimetres (five inches) thick; so less space is required for nesting than one might expect. Nevertheless, the size of the stub or tree necessary for nesting is proportional to the size of the woodpecker. Also, the size and abundance of dead or dying trees influence the abundance and types of woodpeckers that can occupy an area.

From two to nine white eggs are laid on the bare floor of the woodpecker's nest cavity. Most species probably are indeterminate layers. By removing the freshly laid eggs as they appeared daily, one experimenter induced a female flicker to lay a total of 71 eggs. The incubation period of woodpeckers varies from 11 to 20 or more days. Both adults incubate the eggs, and typically the male alone occupies the nest during the night. The young are naked and helpless when hatched and spend up to 36 days in the nest. Both parents feed the young and keep the nest clean by removing excreta frequently. In some species the young are driven off by the adults soon after they can feed themselves; in others the family remains together longer, sometimes even until the next breeding season.

Vocalizations. The voices of Piciformes are rarely melodious and are often harsh or strident. The vocalizations of jacamars are squeaky, the notes sometimes being run together into a trill. Whistles or trills may be alternated or mixed, forming a simple song. Puffbirds are relatively quiet, producing thin whistles, peeps, and twitters. The vocalizations of toucans are loud and often harsh, especially those of the larger species, such as those in the genus *Ramphastos*. The calls of barbets are monotonous and repetitive, but some are bell-like and pleasing to the ear, especially those rendered antiphonally (*i.e.*, the birds alternating) or simultaneously by members of a pair. Displays may accompany this singing. With some colour patterns like their relatives the woodpeckers, barbets have similar displays, such as head swinging, head bobbing, bill pointing, and others.

Although the vocalizations of woodpeckers are individually less complicated than the songs of passeriform birds (the order that includes the songbirds), up to eight or nine different and varying calls may be used by some species (*e.g.*, the ladder-backed woodpecker, *Dendrocopos scalaris*). These function as "advertising songs" (usually repetitive, often harsh, "rattle" or "wick" calls), as aggressive or submissive calls associated with displays, and in other ways.

FORM AND FUNCTION

Adapted for diverse modes of life, piciform birds vary strikingly in their morphology and in their habits. The interpretation of differences is difficult in many cases especially considering that different structural means may achieve the same purpose. Thus, the fly-catching jacamars are exceptionally long-tailed, but the puffbirds, which feed similarly, are generally short-tailed.

Foot specialization

Characteristic of all members of the order is the yoke-toed (zygodactylous) foot in which the outer toe (toe IV) as well as the hind toe (I) points to the rear. This was long considered an adaptation to climbing or perching on the vertical trunks of trees, but recent studies have shown that this type of toe arrangement may really denote a generalized perching foot. Most of the strongly wood-pecking species of the Picidae that are four-toed spread the toes laterally and anteriorly—*i.e.*, up the tree that they are climbing, with no toes pointing to the rear. Three-toed woodpeckers, such as the northern three-toed woodpecker (*Picoides tridactylus*), which have lost the hind toe (hallux, toe I), climb with all three toes facing to the front and sides and swing the inner toe (II) backward when perching on a branch. The hallux tends to be small in those Picidae that have retained it, and its loss has occurred independently in several lines of woodpeckers and piculets.

Just as birds in general are structurally similar because of requirements imposed by flight, woodpeckers are structurally similar because of requirements imposed by wood-pecking habits. The special modifications responsible for their unique abilities are diverse. The tail feathers (especially the central one or two pairs) are stronger in woodpeckers, resisting the wear caused by their use in propping the bird's body as it hammers with the bill. The toe structure and associated arrangement of tendons and leg muscles form a functional complex of features enabling the woodpecker to climb tree trunks and to maintain its position while pecking the tree.

Another functional complex involves the skull, head musculature, bill, tongue, and associated structures of the woodpecker, collectively its food-gathering apparatus. The skull is unusually thick, and, in the most specialized wood-pecking species, it curves inward at the upper base of the bill, instead of meeting it directly, giving a built-in shock absorber. The head musculature is specialized to power the thrusts of the bill and to help absorb the shock caused by pecking. The long, distensible tongue is variously equipped with tiny barbs and coated with a sticky substance provided by often greatly enlarged and modified salivary glands. The horns of the hyoid apparatus responsible for the ability to extend the tongue are tremendously elongated, and when the tongue is retracted, they often extend around the back of the skull, over the top, and (rarely) even onto the bill and into a nostril (see Figure 45). The tongue can be extended into holes made by the bill and can be used both as a probe and as a "spear" (by means of the barbs and sticky mucus) to locate and extract insects far beyond the bill's reach. The bill of more specialized woodpeckers is chisel tipped, broad based, and very sturdy. The nostrils are slitlike, preventing sawdust and wood chips from entering. Other parts of the body may be modified for support or to prevent damage from the shock of pecking.

Drawing by R. Keane

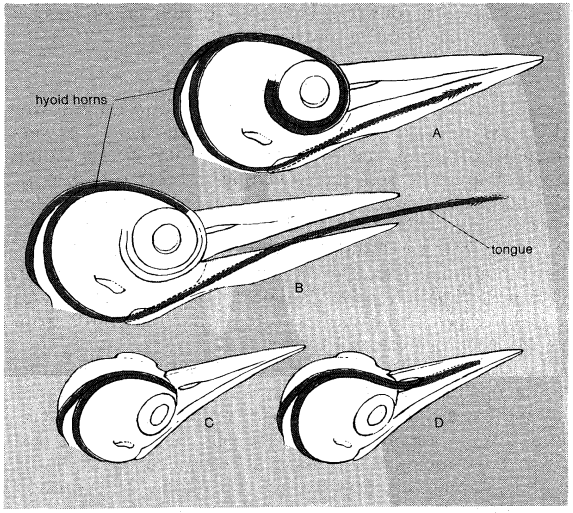


Figure 45: Position of hyoid bones (shaded) with tongue (A) retracted and (B) extended, shown in the hairy woodpecker (*Dendrocopos villosus*). (C and D) Other configurations of hyoids: C ending at the base of the beak; D entering the right nostril and lying within the upper mandible.

An advantageous side effect of their specialization is that woodpeckers are exceptionally tough-skinned and sturdy and thus are difficult to kill. Unfortunately, this has not saved the magnificent, highly specialized ivory-billed woodpecker (*Campephilus principalis*) and its close relative the imperial ivory-billed woodpecker (*Campephilus imperialis*), the largest of all woodpeckers, from extinction through destruction and modification of habitat. A remnant population of the former is known to exist in Cuba, but the future for both species is bleak.

Among woodpeckers, sexual difference in bill structure and other characteristics plays a role in allowing the members of a pair to utilize somewhat different types of food, presumably facilitating more efficient foraging.

Of the other piciform families, the Indicatoridae are specialized in features, such as their thick skin, relating to their beeswax eating and brood parasitism. Other peculiarities, such as the graduated long tail of the lyre-tailed honey guide (*Melichneutes robustus*) and the presence of but nine primaries (there is usually some trace of a tenth primary feather) in the wings of most honey guides, are more difficult to interpret functionally.

The toucan bill

The bill of the toucans, another specialized group, is something of an anomaly. Very large, with serrate edges and usually splashed with colour, it is peculiarly constructed within. Many crisscrossing, fine, bony fibres give it great strength for its size and light weight. The serrated bill may be effective in tearing open fruits, although the birds often feed on small fruits daintily picked, held at the tip of the bill, and thrown back into the mouth. Alternatively, the serrations may aid in securing insects or in plucking young birds out of nests. The colour patterns of the bill probably function as species-recognition signals or in some other way during displays. Yet the size of the bill seems greater than necessary for the serving of these functions. Studies of the feeding habits of the larger species may enable the solution of this point.

EVOLUTION AND PALEONTOLOGY

Fossil species of the order Piciformes are almost unknown. Many avian fossils identified long ago are in need of verification, including the early fossil "woodpeckers," some of which are now believed to belong in other orders. Fossil woodpeckers are now known only from the Pliocene Epoch (about 6,000,000 years ago). Except for a single specimen from the middle Miocene of Bavaria assigned to the Capitonidae, the other Recent families have no fossil records. A fossil family, Zygodactylidae, was erected in 1971 to receive three species of fossil piciform birds from the Miocene of western Europe.

The dearth of fossils leaves the evolution of existing groups open to speculation. If it is true that existing families of the order are only specialized remnants of a formerly more diverse assemblage, then the generalized groups that might have included common ancestors of some of the modern families are now extinct. Therefore, the evolution of the order must be interpreted from inferences based on the morphology of its existing families.

Little can be said about the puffbirds and jacamars. Both families are relatively uniform, and no existing species shows intermediacy or close resemblance between them. They must have evolved very early in the history of the order from a generalized, fly-catching ancestor. If the Galbulae and Pici are indeed closely related, their divergence from a common ancestor must have occurred long ago.

Ancestors of the woodpeckers and barbets

The families within the suborder Pici are definitely interrelated. The honey guides and barbets are morphologically similar, as are the toucans and barbets. The barbets are rather diverse, foliage and branch-foraging piciform birds. An ancient species of barbet, not greatly unlike modern barbets, could have evolved into the ancestral toucan and another such species into the ancestral honey guide. Woodpeckers and barbets share an ancestor probably dating from an earlier period than either the barbet–toucan or the barbet–honey guide ancestor. The peculiar wrynecks, and even the piculets, may represent relicts of early evolutionary radiations in the line leading to woodpeckers. Modern woodpeckers have radiated into many groups, some specializing for an arboreal wood-pecking existence, others remaining generalized, and still others becoming partially adapted to life on the ground.

CLASSIFICATION

Distinguishing taxonomic features. Various characteristics set the Piciformes apart from other orders and distinguish subgroups within this order. Among these features are the condition of the toes and their flexor tendons; the condition of the nostrils; the absence of basipterygoid processes at the base of the skull; the arrangement of the pelvic muscles; the size of the deltoid muscle; the condition of the syringeal muscles.

Annotated classification. The classification below is based generally upon those of a British ornithologist, P.R. Lowe, and an American, C.G. Sibley. The groupings are accepted generally among taxonomists, with disagreement centring about the taxonomic rank of the subcategories (*e.g.,* superfamily versus family) and whether or not to merge part of the order with the perching birds (Passeriformes). The family indicated by the dagger (†) is extinct.

ORDER PICIFORMES

Zygodactylous birds (toes arranged 2 in front and 2 behind); flexor tendons with a vinculum (a connecting strip), digits I, II and IV supplied by flexor hallucis, and flexor digitorum supplies digit III; nostrils impervious and holorhinal (not deeply cleft); eutaxic (5th secondary present), 9 to 11 primaries, 10 to 13 secondaries; 8 to 12 rectrices (tail feathers); aftershaft present; no down in adults; hole-nesting birds, laying all-white eggs. About 370 species.

Suborder Galbulae

Desmognathous (vomer bone of palate small or lacking; maxillopalatines connecting); nude oil gland; 2 carotid arteries; no gallbladder; 10 primaries, 10 to 12 secondaries, 10 to 12 rectrices. Catch food, usually insects, on wing; nest in ground or termite nest.

Family Galbulidae (jacamars)

Tropical and subtropical Central and South America. Bill long. Vomer lacking. Legs with long smooth scales behind and scutellate (small scales) in front. Moderately sized, long-tailed birds with iridescent plumage dominated by green. Feed by fly catching for large insects. Nest in burrows excavated in banks.

†*Family Zygodactylidae*

Middle and upper Miocene of France and Bavaria; 3 species.

Family Bucconidae (puffbirds)

Tropical and subtropical Central and South America. Distinguished by a shorter, wider bill than jacamars, with a decurved tip. Vomer present. Leg scales scutellate in front and behind. Moderate or small-sized, dull-coloured birds with an unusually large head, plumage lax and noniridescent. Feed by fly catching for insects and other arthropods and vertebrates as well. Nest in burrows in the ground, or in termite nests on the ground, or in trees.

Suborder Pici

Rarely desmognathous, usually aegithognathous (vomer broad; maxillopalatines not connecting) or saurognathous (vomer a delicate rod; maxillopalatines reduced, not connecting); oil gland usually feathered; 9 to 11 primaries, 10 to 13 secondaries, 10 to 12 rectrices. Nest in tree cavity, rarely in ground. Food diverse, secured by various means, but rarely by fly catching.

Family Capitonidae (barbets)

Tropical and subtropical New World, Africa, and Asia. Small to medium bill, not chisel-like; short, nonextensible tongue; nostrils opening onto bill surface without a fossa (depression); conspicuous bristles at base of bill. Small to medium, usually brightly coloured birds of stocky appearance. Food fruit or insects, often gleaned from tree bark or foliage. Nest hole in trees or termite nests, usually excavated by the birds.

Family Indicatoridae (honey guides)

Tropical and subtropical Africa and Asia. Small bill, not chisel-like; short, nonextensible tongue; nostrils opening onto bill in centres of large fossae, and without conspicuous bristles at base of bill. Small or medium, generally dull-coloured birds with smallish heads. Food beeswax and insects. Brood parasites, laying eggs in nests of other birds, especially piciform birds.

Family Ramphastidae (toucans)

Tropical to temperate Central and South America. Distinguished by a long and often deep and brightly coloured bill, with serrated edges. Tongue nonextensible but long and fringed. Medium to large, long-tailed and often brightly coloured birds. Food mainly fruit, also insects, even eggs and small birds. Nest in natural tree cavities or woodpecker holes, never excavated.

Family Picidae (woodpeckers, piculets, wrynecks)

Generally worldwide, except Australian region. Distinguished by a straight, usually chisel-like bill, long extensible tongue with elongated hyoid (tongue) apparatus, salivary glands secreting a sticky material that coats the tongue. Very small to large birds, often patterned in black, white, and bright colours. Forage on tree trunks and branches, rarely on the ground. Food chiefly insects, some fruit, rarely sap and seeds. Nest excavated (except wrynecks) in tree, post, cactus, or (rarely) in the ground.

Critical appraisal. Although recognizing the relationship of this order to perching birds (Passeriformes), most taxonomists prefer to keep the two groups separate as distinct orders, especially since clear evidence for relationship with a particular group of perching birds is lacking. No perching bird has the zygodactyl toe arrangement. The exact relationship between the suborders Galbulae and Pici has not been established. Lowe advocated relationship of the Galbulae with the Coraciiformes (kingfishers, bee-eaters, etc.), and the merger of the Pici in the Passeriformes. The suborder Pici is often considered to contain three superfamilies, for the toucans, the woodpeckers, and the barbets and honey guides, respectively. The above arrangement within the suborder Pici begins with the barbets, generally recognized as the least specialized group within the suborder and the most likely ancestral family. The honey guides are almost universally considered very closely related to barbets and are placed next. There is agreement that the specialized toucans, placed next, evolved from barbet-like ancestors. The exact relationship of the specialized woodpecker family with barbets is unknown, but it is likely either that woodpeckers evolved from barbets or that the two groups shared a common ancestor. No authority has suggested that woodpeckers are more closely related to toucans or to honey guides than to barbets. Various authorities have arranged these families diversely, but nearly all place the woodpeckers last. No single arrangement is universally accepted. (L.L.S.)

Passeriformes (perching birds)

Worldwide abundance of passerines

The order Passeriformes, the passerines, or perching birds, is the dominant avian group on Earth today. Considered the most highly evolved of all birds, passerines have undergone an explosive evolutionary radiation in relatively recent geological time and now occur in abundance on all continents except Antarctica and on most oceanic islands. Their rapid evolution and adaptation to virtually all terrestrial environments have resulted in a large number of species, some 5,100, compared to only about 3,500 species for all other birds.

The order Passeriformes is divided by most taxonomists into four suborders: Eurylaimi, Tyranni, Menurae, and Passeres. The first three, containing about 1,100 species, are considered more primitive and are often grouped informally as the "suboscines" for convenient comparison with the very large fourth suborder, the oscines, or songbirds (about 4,000 species).

GENERAL FEATURES

Size range and structural diversity. Passerines are small to medium-sized land birds, ranging from about 7.5 to about 117 centimetres (three to 46 inches) in overall length. Among the tiniest species are the New World flycatchers (Tyrannidae), New Zealand wrens (Xenicidae), titmice (Paridae), flower-peckers (Dicaeidae), tanagers (Thraupidae), and waxbills (Estrildidae). The heaviest are the lyrebirds (Menuridae) of Australia and the ravens (*Corvus*). The longest species, the ribbon-tailed bird of paradise (*Astrapia mayeri*), is actually not so large in body bulk but has extremely long tail feathers. Most passerine species fall within the range of about 12.5 to 20 centimetres (five to eight inches) in length and from 15 to 30 grams (0.5 to one ounce) in weight. A house sparrow (*Passer domesticus*), for example, is 12 to 15 centimetres (five to six inches) long and weighs about 26 grams (0.9 ounce); a cardinal (*Cardinalis cardinalis*) 20 to 23 centimetres (eight to nine inches) and approximately 44 grams (1.5 ounces).

Passerines have evolved a great diversity of feeding adaptations. The majority are insectivorous, at least at certain times of their lives. Members of the order have evolved many ways for finding insect food: swallows (Hirundinidae) are aerial feeders; New World flycatchers "hawk" insects by flying out from a perch; vireos (Vireonidae) glean insects from small twigs and foliage; woodcreepers (Dendrocolaptidae), nuthatches (Sittidae), and creepers (Certhiidae) search for insects in crevices in tree bark; and many other species pick and scratch on the ground and in leaf litter. More specialized passerines eat aquatic insects (dippers: Cinclidae), fish (some New World flycatchers: Tyrannidae), fruit (cotingas: Cotingidae; and many others), leaves (plantcutters: Phytotomidae), nectar (sunbirds: Nectariniidae), small land vertebrates (shrikes: Laniidae), and seeds (finches and many others). For these different food habits, various structural specializations have developed, especially in the bill and feet (see below *Form and function*).

Importance to man. *Aesthetic and economic importance.* Since prehistoric times people have enjoyed watching and listening to songbirds. The almost infinite variety of colours, patterns, behavioral traits, songs, and calls found in these birds appeals to man's aesthetic tastes. As objects of beauty and interest passerines have been incorporated into human culture, folklore, poetry, music, sculpture, and painting. Songbirds have also been used as symbols, such as the European goldfinch (*Carduelis carduelis*), which represented the Passion of Christ in Renaissance art, or the raven (*Corvus corax*), which sometimes has signified a messenger for the devil, an evil omen.

Songbirds as religious symbols

Passerines are widely kept as cage birds. The origins of this practice are lost in antiquity, but it is known that by the 5th century BC the Greeks kept a variety of songbirds, including finches, nightingales and other thrushes, magpies (*Pica*), and starlings (*Sturnidae*). Canaries (*Serinus canaria*) were brought to Europe from their native Canary Islands in the 16th century and have since been developed by domestication and breeding into many varieties. Other passerines now widely kept as pets are the cardueline and estrildine finches and the starlings (particularly Asian mynahs, *Gracula*). The magnitude of the cage-bird "fancy" is indicated by importation statistics on wild and semi-domestic birds: in one recent year alone, over 420,000 passerines (excluding canaries) were legally imported into the United States as cage birds, a number far exceeding that of parrots, the only other bird group whose members are commonly kept as pets. Many countries, including the United States and Great Britain, prohibit the capture and sale of nearly all native songbirds.

Songbirds are also economically important in other ways. Although seldom considered food in economically advanced areas, they are nonetheless important dietary items in many rural or heavily populated countries. China, Japan, and other Oriental countries, for instance, have highly developed techniques for catching small birds, and in cities such as Hong Kong and Tokyo passerines are commonly sold in food markets. In Italy, France, and Belgium the capture of migratory songbirds for the pot or for cage birds is still extensive. Laws against such activities are difficult to enact or enforce in areas in which the habit has become part of the culture; hence, for example, more than 10,000,000 passerines were still being caught annually in Italy in the late 1960s.

Fortunately, killing songbirds for their feathers is no longer as prevalent as it once was. Until the early 20th century, however, there were few protective laws, and the wearing of birds (especially on women's hats) was common. In 1886 a young ornithologist reported that he had counted feathers from no fewer than 40 bird species, including 22 kinds of passerines, on hats seen on two afternoon walks in a fashionable part of New York City.

Other cultures have used songbird feathers for personal adornment, but usually for men rather than women. This practice often came about not only for the beauty of the feathers themselves but also because the feathers were used as symbols of such bird qualities as speed and aggressiveness. Most noteworthy are the feathers of male birds of paradise (Paradiseidae) used as headdresses by tribesmen of New Guinea. An estimated 80,000 adult birds are still being killed annually for this purpose. Other ancient uses of passerine feathers have now largely been terminated, either because the birds are extinct (in the case of Hawaiian feather cloaks) or because more suitable modern substitutes have been found (Melanesian feather money).

Songbirds as agricultural pests

Some passerines, on the other hand, are serious economic pests. In areas in which one-crop agriculture is extensive, certain bird species have undergone population explosions because of almost unlimited food availability; in turn, their crop depredations can be serious. One example of this is in Africa, where immense flocks of a small weaver, the red-billed quelea, or Sudan dioch (*Quelea quelea*), numbering as many as 20,000,000 birds in one flock, do millions of dollars worth of damage to various small grain crops each year. Other serious pests are the Java sparrow (*Padda oryzivora*) in Asian rice fields and mixed flocks of New World blackbirds (*Icteridae*) and European starlings (*Sturnus vulgaris*) in grainfields in the United States. The same starling and the house sparrow, both introduced to the United States from Europe, have become urban pests by fouling buildings with excrement and blocking rain gutters and ventilators with their nests. Starlings occasionally also have been implicated in accidents; in 1960 a flock at the airport in Boston was sucked into a jet's engines and the resultant crash killed 61 people.

Ecological importance. The greatest importance of passerines is ecological. As the dominant form of birdlife in virtually all terrestrial environments, the perching birds are a major component of the world's ecosystems. They consume great quantities and varieties of food—grains, fruits, insects and other invertebrates, small amphibians and reptiles, and even small mammals—and in turn serve as food for other animals; they act as hosts for parasites and are occasionally parasitic themselves; they both propagate and distribute plants by pollinating flowers and carrying viable seeds to new locations; and they have the mobility (through migration) to utilize habitats that are available only at certain times of the year. A few aspects of the ecological impact of passerines are known, but, until the science of ecology has advanced, the true magnitude of their importance cannot be evaluated with precision.

NATURAL HISTORY

Types of territories

Reproduction. *Territoriality and courtship.* The breeding behaviour of passerines is diverse. Most species are solitary nesters, a single monogamous pair of birds maintaining a territory that is large enough to support all their activities during the breeding season: courtship, mating, nesting, and food gathering. Others have similar territories, but the birds forage outside the defended area for most of their food (*e.g.,* the North American redwinged blackbird, *Agelaius phoeniceus*). Still others are colonial nesters, defending only the nest site and a small area immediately adjacent to it. Some species build individual nests close together in a colony (oropendolas, Icteridae; some swallows; the house sparrow), and others construct massive communal nests in which the breeding pair defends only its own nest cavity (palm chat, *Dulus;* several weaverbirds, Ploceidae). In a few species polygynous (polygamous) males establish special display territories (leks) for courtship and mating in which no nesting takes place. In these courtship arenas the males, usually brilliantly coloured, attract females through song and posturing and sometimes by dancing, manipulation of objects, and other elaborate displays. The best known arena-displaying males are the cocks of the rock (*Rupicola*), manakins (Pipridae), birds of paradise, and bower-birds (Ptilonorhynchidae). After mating in or near the lek, a female leaves to build a nest and raise the young without assistance from the male. Still other species build no nest at all, but are brood parasites (some cowbirds, Icteridae; whydahs, Ploceidae); the female lays her eggs in the nests of other (usually smaller) species, and the young are raised entirely by the foster parents.

Nesting. Nest sites are varied: they include holes in the ground, trees, banks, and rock crevices; on ledges; on the surface of the ground; within the larger nests of other species (including nonpasserines) or near wasp nests (presumably for the protection the wasps afford); and in a wide variety of vegetation—grasses, shrubs, and trees.

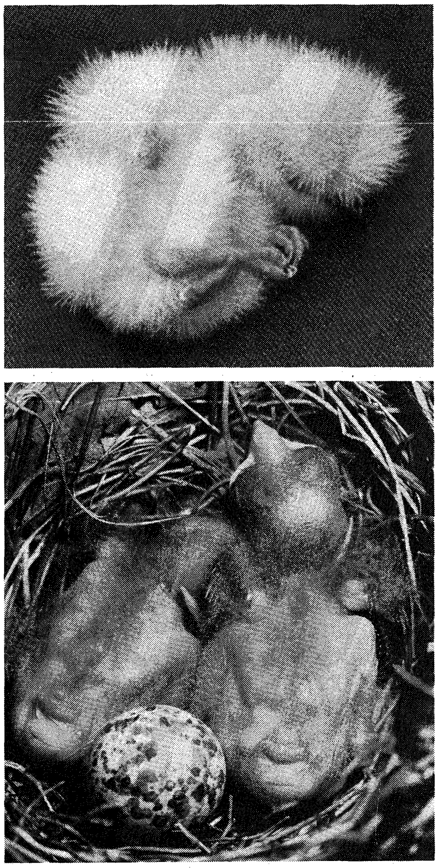
Nest materials

Passerine nests are usually elaborately constructed and may contain many different kinds of materials: mud, grasses, hair and feathers, strips of bark, plant fibres and downs, rootlets, twigs and sticks, leaves, string, spider webs, cast snake skins, lichens, and many other substances. Most species build open nests, usually cup-shaped. Others form domed or ball-shaped closed nests, with an entrance at the side (occasionally at the top or bottom). One of the most famous closed nests is that of the South American ovenbirds of the genus *Furnarius* (Furnariidae), whose name derives from its thick-walled mud "oven" nest, often built on top of a fence post or some other exposed site. The North American ovenbird, *Seiurus aurocapillus* (a wood warbler, Parulidae), also builds a domed oven-shaped nest, but of plant materials on the forest floor. Some species, especially members of the Icteridae, make soft hanging nests that range to two feet or more in length. The thorn birds (*Phacellodomus*), as well as many other Furnariidae, build huge nests of twigs suspended from the ends of tree branches; these nests, which may be more than two metres (nearly seven feet) long and contain many compartments, are used by only a single nesting pair, sometimes with nonbreeding helpers (probably the young of the previous season). These nests are often appropriated by troupials (*Icterus icterus*), which evict the owners, even destroying the eggs and young in the process. a few other species also take over nests for their own use, notably the piratic flycatcher (*Legatus leucophaius,* a tyrannid) and the bay-winged cowbird (*Molothrus badius*).

Woven nests

Nests of many passerines are constructed with amazing skill. The tailorbirds of Asia (*Orthotomus*) are noted for nests built in a pocket that the birds make by sewing together the edges of one or more leaves, using plant fibres or other materials. Some species, especially the weavers, are able to tie knots with strips of grass or palm leaves and thus weave an exceptionally tight and compact nest. Others build equally firm nests by felting the materials together. In contrast, a few passerines build flimsy nests (some Cotingidae), apparently as an adaptation toward lessened visibility to predators, for such nests are attended minimally by the parents, seemingly to draw as little attention to the site as possible. Other birds excavate their nests in soft earthen banks, use old woodpecker holes, or find natural crevices in trees or rocks. The type of nest built by the members of a single family may be varied (extremely so in the Furnariidae) or consistent: all woodcreepers nest in holes; all vireos weave a cup between the arms of a forked branch.

(Top) David W. Snow, (bottom) Root Resources—EB Inc.



Two main types of passerine young.
(Top) Downy young of the South American bare-throated bellbird (*Procnias nudicollis*). (Bottom) Naked young of the North American cardinal (*Cardinalis cardinalis*).

Incubation and parental care. Passerines lay clutches of one to 14 eggs, clutch size being unrelated to the size of the bird. The largest species, the two lyrebirds (*Menura*), lay a single egg; some of the smaller titmice (*Parus*) have been recorded with the biggest clutches. In most passerines the female incubates the eggs alone, but in some groups—such as the antbirds (Formicariidae), certain grosbeaks (*Pheucticus*), and others—the male shares equally in incubation. Males of most species help to feed the young. some passerines have only one nest per breeding season, but others may have two or more, especially if one nest is destroyed before the young fledge. The incubation period generally varies from 11 to 21 days depending on the species but is well over a month in lyrebirds. The hatchlings are typically blind, sparsely covered with down, and helpless; some species hatch completely naked, and a very few are densely covered with down at hatching (some cotingas, antbirds of the genus *Formicarius,* and some Campephagidae). The young remain in the nest for eight to 30 or 35 days (about 42 in the lyrebirds) but most commonly from 10 to 15 days. After they fledge, they require some days or weeks to become fully independent of their parents.

Variety of vocalizations among passerines

Sound production. An outstanding aspect of passerine behaviour is the ability to sing. Song is best developed in the oscines, which have a highly complex vocal organ or syrinx, but even the more primitive suboscines are capable of a variety of vocal sounds. The woodcreepers (Dendrocolaptidae), ovenbirds (Furnariidae), and antbirds (Formicariidae) sing relatively simple songs, consisting of a few notes or whistles, often repeated rapidly in a trill, roll, or rattle. Manakins (Pipridae) also utter simple trills or whistles: in addition, some species are capable of a loud nonvocal snapping sound, which is produced by specialized wing feathers. The cotingas sing a wider variety of songs, from quiet musical notes to the incredibly loud and far-carrying "gongs" of the bellbirds (*Procnias*). The New World flycatchers are well known for their range of distinctive call notes, and many species sing well and melodiously. In some groups (notably the *Empidonax* complex), the plumages of closely related species are so similar that the birds can be distinguished in the field only by their calls and songs. Both the lyrebirds and scrubbirds (Atrichornithidae) have syrinxes more like those of oscines and are known for their loud and complicated songs. They are also accomplished mimics; lyrebirds mimic the songs of almost all birds within their hearing, as well as many mechanical sounds. Many species of oscines have complicated and beautiful songs, notable examples being the nightingales (*Luscinia*) and some other thrushes, larks (Alaudidae), mimic thrushes (Mimidae), and wrens (Troglodytidae). The possession of the complex oscine syrinx does not guarantee a complex song, however, and many "songbirds," such as waxwings (*Bombycilla*) and swallows, utter simpler sounds than do many suboscines.

Only the male of most passerine species sings a true song, although the female can produce a variety of call notes and other sounds. In some species in which the female sings, she seldom does so during the breeding season unless it is a duet with her mate. Such duetting or antiphonal singing of paired birds is so well developed in certain species that it is difficult to determine that the song is coming from two individuals. In the African black-and-red shrike (*Laniarius barbarus erythrogaster*) the reaction time between the male's start of song and the female's response has been timed at 0.135 second.

Active versus passive anting

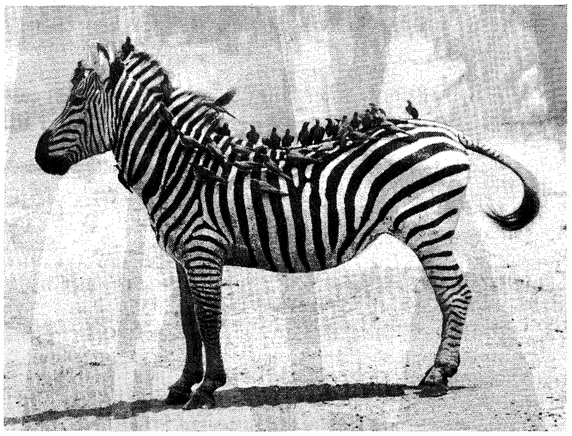
Interactions with ants. *Anting.* A characteristic but poorly understood behaviour pattern of passerines is the practice of anting. This peculiar ritual has two forms: active anting, in which a bird picks up worker ants in its bill and wipes them on its feathers in a stereotyped manner, and passive anting, in which the bird squats or lies down in a group of ants and assumes an exposing stance so that the ants will crawl up into its feathers. Birds may also apply ants to their plumage while passively anting, but species that use the active stance (the majority of recorded passerines) apparently never use the passive stance. Birds show definite discrimination in the type of ants used, avoiding stinging species and selecting those that exude or spray formic acid or other defense fluids (ants of the subfamilies Formicinae and Dolichoderinae of the family Formicidae). A great deal of controversy has existed over the function of anting. Some authorities have theorized that it is a form of self-stimulation, but British ornithologist K.E.L. Simmons has argued convincingly that it is a type of feather maintenance. Formic acid and other ant fluids are known to be insecticidal; dressing the feathers with ants would thus kill or deter avian parasites, such as lice and mites. Additional components of ant fluids include essential oils, which could be used by birds to supplement the oils from their own uropygial (preen) gland. After a bout of anting, birds often continue feather-maintenance activities by bathing, oiling (from the uropygial gland), and preening. Recent studies have shown anting to be most prevalent during molt, when the bird's skin is irritated by the growth of new feathers. Anting clearly is innate behaviour, and its remarkable uniformity in at least 30 passerine families, both oscine and suboscine, implies that it has real importance to the bird. Some individuals have been seen to ant with such things as cigarette butts, orange peels, mothballs, and smoke, apparently reacting to the pungent fumes of these objects as to the strong odours of ants. A few nonpasserines have also been observed going through motions that are similar to anting, but as yet true anting is known only in the Passeriformes.

Ant-following. Another specialized form of behaviour, also associated with ants, is the practice known as antfollowing. In the New World tropics, nomadic army ants move in huge troops, swarming over the forest floor in columns as wide as 10 metres (more than 30 feet) or more. Because the ants devour all the small animal life in their path, a moving column of them is edged by fleeing insects, spiders, millipedes, isopods, small frogs, and lizards. The ant columns are accompanied by troops of birds that seize the fugitives. Ant-following birds apparently do not eat the ants but only the insects and other small animals trying to escape. A number of passerine species, notably several antbirds, are believed to be entirely dependent on army ants for finding food. Many other birds also follow ants when they come upon them; these include woodcreepers, manakins, New World flycatchers, tanagers, wrens, and occasional ovenbirds. Even some nonpasserines may join a troop of ant-followers—motmots (Momotidae), tinamous (Tinamidae), and hawks—although the hawks may be more attracted by the ant-following birds than by the insects. The same ant-dependent species have also been known to follow large animals, including man, that stir up insects with their feet.

Association of passerines with large animals

A few passerines, although not ant followers, will escort large quadrupeds, such as cattle, buffalo, and deer, to catch the insects that fly up around them and to feed on the ticks and flies parasitizing the animals themselves; especially noted for this behaviour are the cattle tyrant (*Machetornis rixosa,* Tyrannidae) tickbirds or oxpeckers (*Buphagus,* Sturnidae), and several cowbirds. In Australia yellow robins (*Eopsaltria,* Muscicapidae) will follow the much larger lyrebirds as they scratch and feed along the ground.

Popperfoto, London



A zebra covered by a large number of oxpeckers.

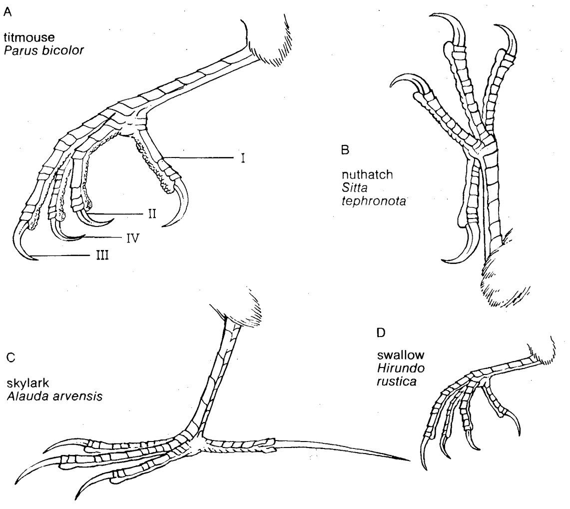


Figure 46: Modifications of the foot of perching birds, for (A) perching and clambering; (B) bark climbing; (C) ground walking; (D) perching only (weak foot). Right feet are shown.

Drawing by R. Keane

FORM AND FUNCTION

External features. *Feet and legs.* The single feature that distinguishes passerines from all similar birds is their "perching" foot. In this foot type, all four toes are well developed and free from one another; in some families (wrens and most suboscines) the front toes may be partially fused at the base, but the distal portions (extremities) are functionally free. The hind toe (hallux) is joined on the same level with the front toes and opposes them, so that the foot can grip a perch. The only exception to this passerine foot type is found in the well-named *Paradoxornis paradoxus,* or three-toed parrotbill (Panuridae), in which the outer toe is reduced to a short clawless stump, fused to the middle toe; other species of *Paradoxornis* have normal feet.

Modifications of the foot

Although all passerines can perch, not all do so habitually. A number of species (some tapaculos, Rhinocryptidae; larks; pipits, Motacillidae) are largely terrestrial and have feet modified for walking and running; the terrestrial foot is differently proportioned from the typical perching one, often with longer toes and longer, straighter claws (particularly on the hallux), probably as an aid in maintaining balance when running. The dippers, or water ouzels (*Cinclus*), are semi-aquatic, but although they successfully swim on the water surface and walk underwater searching for food on stream bottoms, they have retained the typical passerine foot. The single slight difference in the *Cinclus* foot is that the claw of the middle toe sometimes has a thin horny flap (of unknown function) on its inner border. Some other passerines, notably swallows, live a largely aerial life and have small and weak feet. The typical arboreal songbird has a well-developed foot, with the middle front toe longer than the others. Birds such as woodcreepers and nuthatches that often cling to vertical surfaces have strong, curved, sharp claws. Those that spend much of their time walking and scratching on the ground (although not limited to terrestrial activity) tend to have heavy, straighter, and rather blunt claws. Most passerines, however, have moderately curved sharp claws that are suited to grip a variety of rounded or rough surfaces.

The lower leg of passerines, the tarsometatarsus (usually called simply the tarsus), is normally covered by a horny sheath (podotheca). Exceptions include some swallows, which have feathered tarsi. Although the various different patterns of scale size and distribution of the normal unfeathered podotheca have been used by some taxonomists to differentiate families or groups of families, study has revealed so much variability in the tarsal patterns of certain families that it is no longer considered a reliable family character; it may still be useful as a generic or specific character. In most oscines the posterior (plantar) surface of the tarsus is bilaminate—*i.e.,* covered by two long plates, or laminae.

Evolutionary plasticity of the bill

Bill. The bills of passerines are extraordinarily diverse in size, shape, and proportions. This diversity was long thought to be indicative of the birds' relationships and so was used as a prime taxonomic character. It is now believed, however, that bills are evolutionarily plastic, reacting with relative ease to selective pressures, particularly to changes in feeding habits. Thus on a broad scale a passerine's bill shape reveals less about its family affinities than it does about its food preferences, and, although bill shape may be an aid to determining a bird's relationships, it must be considered in the light of other features and of the degree of variation found in the family. Two frequently cited examples of the adaptiveness of bills are the geospizine, or Darwin's finches (Fringillidae) of the Galápagos Islands, and the Hawaiian honeycreepers, Drepanididae (see EVOLUTION, THE THEORY OF). Each is a closely interrelated group of birds with different kinds of bills in the several species and genera. Those of drepanidids range from heavy, seedcracking, grosbeak-like bills through thin, pointed insectivorous types to the long, decurved (curved downward) bills of nectar feeders. These Hawaiian birds are now thought to be members of a single family of nine genera. On the basis largely of bill shape, they were once classified into four different families and 18 genera.

Most birds, including passerines, show little sexual dimorphism (difference between sexes) in bills except for minor differences in size (reflecting general body size differences) and sometimes in colour. The most outstanding exception is the extinct huia (*Heteralocha acutirostris,* Callaeidae), originally classified as two different species. The male of this New Zealand bird had a strong, chiselling bill, whereas the female had a long, decurved, pliable bill. Reportedly the two sexes fed cooperatively, the male digging in decaying wood and the female probing in crevices to extract grubs. The species unfortunately was prized by the Maoris, who used the white-tipped tail feathers in ceremonial headdresses, as well as by Europeans, and, after most of its habitat had been destroyed, the huia was hunted to extinction about the end of the 19th century.

Drawing by R. Keane

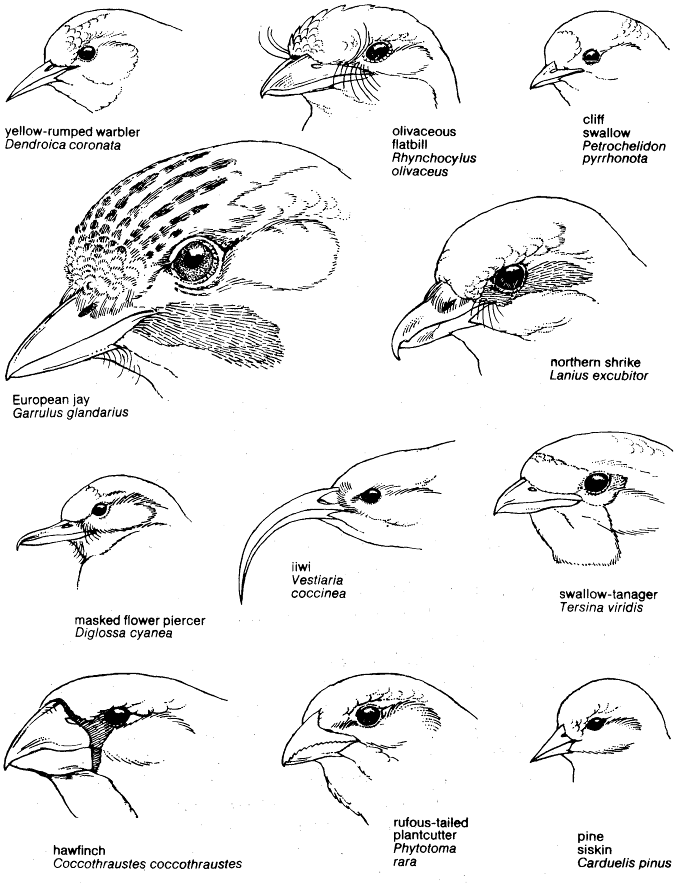


Figure 47: Types of bills found among passerine birds.

Specializations of the bill

Passerine bills may be broadly classified into eight morphological and functional types:

1. Insectivorous: a generalized type found in many passerines, ranging from relatively straight and pointed (as in the wood warblers, Parulidae), through bills with a slight or pronounced hook (some New World flycatchers), to those that are short, with a wide gape and usually surrounded by rictal bristles (stiff hairlike feathers)—*e.g.,* in aerial feeders, such as swallows. Most insectivorous bills are relatively light in build, but this depends on the type of insect usually taken by the species and also on how generalized a feeder it is.

2. Omnivorous: unspecialized in shape and function but usually strongly built, as in crows and jays (Corvidae).

3. Toothed: strongly hooked at the tip and with a "tooth" (notch) on either tomium (cutting edge) of the upper mandible; adapted to tearing up large, relatively soft prey. This is the typical bill of shrikes (Laniidae) but is also found in some unrelated birds, such as the Australian bellmagpies (Cracticidae) and some tanagers.

4. Tearing: a relatively light bill with a strong hook at the tip, for tearing open objects, such as flowers, to obtain the insects and nectar within. Found in flower piercers (*Diglossa,* Thraupidae).

5. Probing: relatively narrow and often downcurved; slender in species that probe flowers for tiny insects and nectar (sunbirds; some Hawaiian honeycreepers) but more heavily constructed in those that probe in wood or under tree bark (creepers, *Certhia;* some woodcreepers).

6. Frugivorous: variable but usually rather wide; ranges from lightly built with a wide gape for swallowing whole fruits (found in some cotingas, and in the swallow-tanager, *Tersina*) to more heavily built for tearing apart tougher fruits (some tanagers).

7. Serrated: conical, with a finely serrated edge, adapted for feeding on leaves, buds, shoots, and fruit. Found only in the plant cutters (Phytotomidae).

8. Conical: adapted for seed eating. Ranges from exceedingly stout and blunt (*e.g.,* the hawfinch, *Coccothraustes,* which can crack remarkably hard objects, such as cherry pits) to relatively small and pointed (siskins, *Carduelis*). Some forms specialized for particular kinds of seed extraction (*e.g.,* crossbills, *Loxia,* which feed on pine seeds).

Opportunistic feeding

This classification indicates morphological and functional types of bills, but it does not imply that a species with a particular type of bill will feed only on the food for which it is best adapted. Although some birds are extremely specialized in their feeding habits, most are opportunistic feeders, seizing upon whatever food is readily available and can be "handled" with the bill. Hence many basically granivorous or frugivorous birds catch insects, especially when feeding nestlings, and many insectivorous species exploit seasonally available plant food. Myrtle warblers (*Dendroica coronata*) and tree swallows (*Iridoprocne bicolor*), for example, feed on bayberries in fall and winter, and eastern kingbirds (*Tyrannus tyrannus*) and other New World flycatchers will eat a variety of fruits and berries in season.

The mandibles of passerines, like those of all other birds, are composed of bone covered with a horny sheath, the ramphotheca. The ramphotheca is worn down by normal use and, in most birds, is capable of growing to replace the lost material. In individuals with damaged bills or those (such as cage birds) that do not have the opportunity to wear down the constantly growing ramphotheca, the bills overgrow at the tip.

Plumage and pterylosis. The colours, patterns, and textures of passerine feathers are considered important taxonomic characters, especially in determining genera, species, and subspecies. Plumage is also occasionally used in a very broad way to indicate evolutionary levels. Spots, streaks, and dull colours are generally considered more primitive than bold or complicated patterns and bright colours, but there are many exceptions to this rule.

Passerines often are sexually dimorphic in their plumage, with adult males wearing brighter colours and more striking patterns than do females. In some families, notably tanagers (Thraupidae), wood warblers (Parulidae), and New World orioles (Icteridae), the temperate zone species show more sexual dimorphism than do tropical members of the same families. In addition, many species (especially those in temperate climates) are seasonally dimorphic, with a bright plumage during the breeding season and a dull one in winter. Juvenile plumages of both sexes tend to be cryptically coloured (*i.e.,* adapted for concealment), as is that of the adult female.

Colour varieties in passerines

Virtually any colour may be found in one passerine or another, and the order offers a wide array of specialized feather types, such as the waxlike tips on the flight feathers of waxwings (Bombycillidae); the tufts of stiff feathers in some honeyeaters (Meliphagidae); iridescent "spangles" in some manakins, sunbirds, and tanagers; and the almost unbelievable array of "wires," iridescent gorgets, velvety ruffs, racquet tails, and filamentous plumes of the birds of paradise.

Another taxonomically important character is the number and distribution of feathers (pterylosis) on the bodies of passerines. From external appearance all birds seem to be more or less evenly covered by feathers; in actual fact, however, most birds have their feathers growing from relatively narrow tracts (pterylae) in the skin. From the pterylae the feathers fan out and cover the remainder of the bird's body. In passerines, the feathers are arranged in eight distinguishable tracts, with apteria (relatively bare skin) between them. Variations in tract width and length and especially differences in feather number and distribution are often useful in determining relationships. Of particular interest are the occurrence of apteria within tracts and the configuration of the ventral tract. Also used in classification are the numbers of flight feathers. The remiges (flight feathers on the wings) of most passerines consist of 10 primaries on the "hand" (manus) and nine secondaries on the forearm (ulna). In all perching birds the tenth (outermost) primary is reduced to some degree, and in many families only nine may be found. The number of secondaries is more variable, with some species having as many as 14 (the satin bowerbird, *Ptilonorhynchus violaceus*). Tail feathers (rectrices) also vary; most passerines have 12, but the number ranges from six to 16.

Of importance in some species is the relative length of the primaries. This "wing formula" is often useful in distinguishing between species of such difficult groups as the New World flycatchers and the Old World warblers (Sylviidae).

Internal features. *Syrinx.* In a group of birds as vocal as the passerines, it is natural that the structure of the vocal apparatus should have evolutionary significance. Differing from the mammalian larynx in both location and structure, the syrinx consists of a resonating chamber at the lower end of the windpipe (trachea), with associated membranes, cartilages, and muscles. These modifications involve elements of the bronchi (the two tubes connecting the trachea with the lungs) as well as those of the trachea. Since the mid-19th century the basic subdivisions of the order Passeriformes have been based primarily on the structure of the syrinx. Some of the features of the syrinx long thought to be indicators of relationships have recently been shown to have little taxonomic validity. Other taxonomic characters now must be studied and evaluated before the classification can be adjusted.

Syringeal muscles are classified into two groups: extrinsic muscles, which connect the syrinx with other parts of the anatomy, and intrinsic muscles, which extend from one part of the syrinx to another. The number, shape, and attachments of the intrinsic muscles are likely to remain important in passerine classification. Those birds in which the muscles are inserted on the middle of the bronchial semi-rings (C-shaped cartilages that strengthen the bronchi) are sometimes called mesomyodian (most members of the suborder Tyranni) and those with the insertion on the ends of the semi-rings are acromyodian (Menurae, Passeres). The Eurylaimi and a few others have no intrinsic muscles. Further distinction is made in the number of pairs of intrinsic muscles, most importantly in the Passeres, which have four.

Vocal-organ specializations

The passerine syrinx exists in four basic types:

1. Unspecialized: relatively little modification of the tracheobronchial region; few, if any, cartilaginous spe-

cializations, and no intrinsic muscles; found in broadbills (Eurylaimidae), pittas (Pittidae), New Zealand wrens, asities (Philepittidae), plantcutters, most cotingas, and a few manakins and tyrant flycatchers.

2. Tracheophone: most of the specializations limited to the tracheal region; intrinsic muscles number zero to two pairs; pessulus (a bony bar lying at the junction of the bronchi) absent; found in all members of the Furnarioidea (South American ovenbirds, woodcreepers, antbirds, and tapaculos).

3. Intermediate tracheobronchial: various modifications of cartilages and membranes; one or two pairs of intrinsic muscles; pessulus present or absent; found in the sharpbill (*Oxyruncus*) and most manakins and tyrant flycatchers.

4. Oscine (acromyodean): complex musculature involving four pairs of intrinsic muscles (but three pairs in lyrebirds and scrubbirds); some cartilaginous specializations; pessulus present (except in larks).

Despite the rather extensive knowledge of the morphology of the syrinx and its importance in passerine classification, very little is known about how it functions to produce sounds. Several theories have been proposed, but none has yet been proved wholly satisfactory.

Skeleton. Of the many variations in passerine skeletal structure, only a few that are important in classification are mentioned here.

In the skull the bony palate, composed of a number of small bones, is termed aegithognathous; also found in swifts (Apodiformes), this palatal type is characterized by the shape and type of fusion of the small bones of the palate. Within this basic type the many minor variations in shape, size, and position of the component bones are useful in delimiting closely related groups of birds, especially suboscines.

Elsewhere on the head, variations in the hyoid apparatus, a complex of small bones that supports the tongue, have been used in passerine classification.

In the sternum (breastbone) the shape of the anteriormost spine (spina sternalis) and the number of notches in the posterior border are of great interest. The spina sternalis, which is short and forked in most passerines, is long and simple in the Eurylaimidae (one exception), the Philepittidae, and a few of the Cotingidae. All oscines and most suboscines have a single pair of posterior sternal notches; only the tapaculos and certain of the terrestrial antbirds (*Conopophaga, Pittasoma, Hylopezus, Myrmothera*) have two pairs. The sternum of lyrebirds differs from those of all others in the order in being very thick, long, and narrow; it may have no posterior notches at all, or it may have a single shallow pair.

From P.L. Ames in (A,B,C) *Bulletin of the Peabody Museum of Natural History*, no. 37 (1971), Yale University; (D) J.C. George and A.J. Berger, *Avian Myology* (1966), Academic Press

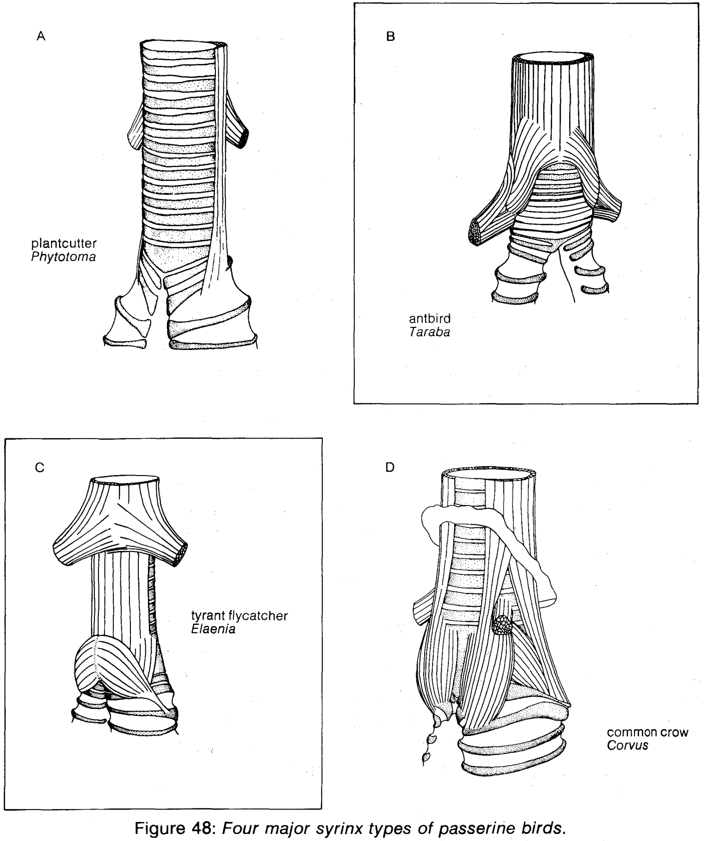


Figure 48: *Four major syrinx types of passerine birds.* Figures show lower trachea, upper bronchi, and attached muscles, seen from the ventral side. (A) unspecialized; (B) tracheophone; (C) intermediate tracheobronchial; (D) oscine.

Musculature. A number of different muscle systems have been important in passerine classification. Important examples, in addition to those of the syrinx, are the muscle complexes controlling the tongue, the jaws, the wings and pectoral girdle, and the legs and pelvic girdle. One character that has been used for over a century is the condition of the deep plantar tendons. These narrow straps extend from the bellies of the two deep flexor muscles on the leg, down the back of the tarsometatarsus, and attach to the toes. They act to close the toes (hence to grasp a perch). In the Eurylaimidae these tendons are connected by a short band (vinculum), but in all other passerines they are entirely separate. This difference has been used by some to divide the passerines into two major groups: the Desmodactyli (vinculum present) and the Eleutherodactyli (vinculum absent).

EVOLUTION AND PALEONTOLOGY

Origins of perching birds

Passerines, like other small land birds with fragile bones and arboreal habits, are rarely fossilized; hence the fossil record gives few clues to their evolution. Their origins, both in time and ancestral type, can only be guessed. Some authorities believe that passerines arose during the Early Cretaceous Era (about 120,000,000 years ago); others believe that it was much later, not until the Late Cretaceous or even Paleocene Epoch (perhaps 65,000,000 years ago).

The earliest known passerine fossils are from the upper Eocene Epoch (about 40,000,000 years ago). One of these is an extinct genus assigned to the modern family of tapaculos (Rhinocryptidae), another is thought to be related to the titmice (Paridae), and still another to the starlings (Sturnidae). *Palaeospiza,* a primitive oscine classified as a separate family (Palaeospizidae) between the larks and the swallows, is known only from the Colorado Oligocene Epoch (about 35,000,000 years ago). As drying conditions during the Miocene Epoch (26,000,000 to 7,000,000 years ago) reduced the forests and encouraged the spread of grasslands, bird families exploited the newly appearing drier and less forested habitats and radiated into them. There for the first time appear remains of crows (Corvidae), thrushes (Turdidae), wagtails (Motacillidae), shrikes (Laniidae), and wood warblers (Parulidae).

During the Pliocene Epoch (from 7,000,000 to 2,500,000 years ago) the warm, dry conditions continued, and most paleornithologists now believe that all of the living passerine families were in existence by its close. They also believe that most modern species of birds arose during the Pleistocene Epoch (about 2,500,000 to 10,000 years ago), a period of almost constant change during which there were four major advances and retreats of the glaciers. Most of the passerines in the fossil record are from the Pleistocene or Recent epochs and represent either living species or close relatives. Evolution since the retreat of the last ice sheet (about 11,000 years ago) has been mainly at the subspecies level.

With the lack of a significant fossil record, nothing is known of the type or types of birds from which passerines arose. Studies of the anatomy of modern forms have led to a general agreement that the perching birds developed from more than one ancestral type (polyphyletic origin). But until more evidence is unearthed, nothing further can be said.

CLASSIFICATION

Distinguishing taxonomic features. The classification of passerines is probably the single most vexing problem in avian taxonomy today. Few of the subgroupings (especially oscine families) have strikingly different characteristics,

Suborders Eurylaimi and Tyranni


black-and-yellow broadbill
Eurylaimus ochromalus
Eurylaimidae


straight-billed woodcreeper
Xiphorhynchus picus
Dendrocolaptidae


pale-legged ovenbird (hornero) with nest
Furnarius leucopus
Furnariidae


pearled treerunner
Margarornis squamiger
Furnariidae


black-crested antshrike
Sakesphorus canadensis
Formicariidae


rufous antpitta
Grallaria rufula
Formicariidae


purple-breasted cotinga
Cotinga cotinga
Cotingidae


saffron-crested tyrant-manakin
Neopelma chrysocephalum
Pipridae


blue-backed manakin
Chiroxiphia pareola
Pipridae


yellow-billed tit-tyrant
Anairetes flavirostris
Tyrannidae


d'Orbigny's chat-tyrant
Ochthoeca oenanthoides
Tyrannidae


rusty-backed antwren
Formicivora rufa
Formicariidae


suiriri flycatcher
Suiriri suiriri
Tyrannidae


rusty-margined flycatcher
Myiozetetes cayanensis
Tyrannidae


chucao tapaculo
Scelorchilus rubecula
Rhinocryptidae

Painting by R. Keane

Plate 2 **Birds**

Suborders Menurae and Passeres

king bird of paradise
Cicinnurus regius
Paradisaeidae

Albert lyrebird (1 metre long)
Menura alberti
Menuridae

golden swallow
Kalochelidon euchrysea
Hirundinidae

bronzed drongo
Dicrurus aeneus
Dicruridae

western black-headed oriole
Oriolus brachyrhynchus
Oriolidae

rufous treecreeper
Climacteris rufa
Climacteridae

black-backed butcherbird
Cracticus mentalis
Cracticidae

green jay
Cyanocorax yncas
Corvidae

Nepal treecreeper
Certhia nipalensis
Certhiidae

magnificent bird of paradise
Diphyllodes magnificus
Paradisaeidae

red-headed tit
Aegithalos concinnus
Paridae

orange-winged sitella
Neositta chrysoptera
Sittidae

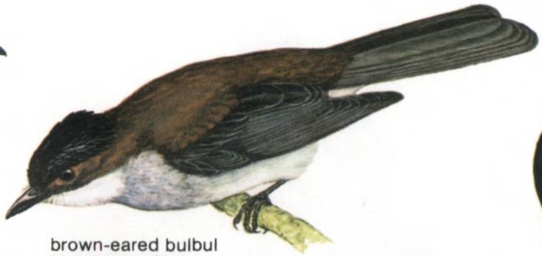
flappet lark
Mirafra rufocinnamomea
Alaudidae

saddleback
Philesturnus carunculatus
Callaeidae

lowland rail-babbler
Ptilorrhoa caerulescens
Timaliidae

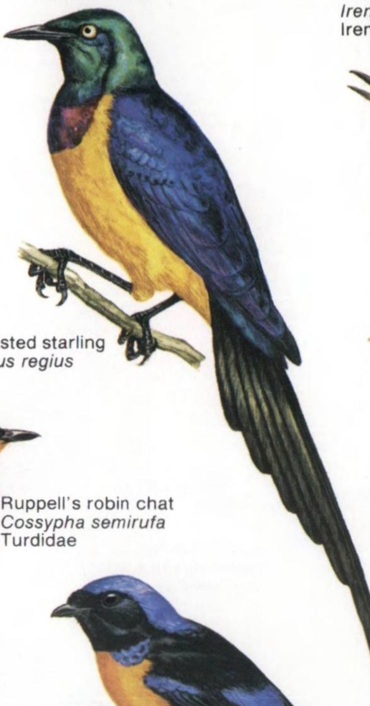
Painting by R. Keane


varied triller
Lalage leucomela
Campephagidae


brown-eared bulbul
Hypsipetes flavala
Pycnonotidae


Philippine fairy bluebird
Irena cyanogaster
Irenidae


black-capped mockingthrush
Donacobius atricapillus
Mimidae


golden-breasted starling
Cosmopsarus regius
Sturnidae


black-bellied wren
Thryothorus fasciatoventris
Troglodytidae


Ruppell's robin chat
Cossypha semirufa
Turdidae


golden-faced pachycare
Pachycare flavogrisea
Pachycephalidae


rufous-bellied niltava
Muscicapa sundara
Muscicapidae

Suborder Passeres


long-tailed silky flycatcher
Ptilogonys caudatus
Ptilogonatidae


white-breasted wood swallow
Artamus leucorhynchus
Artamidae

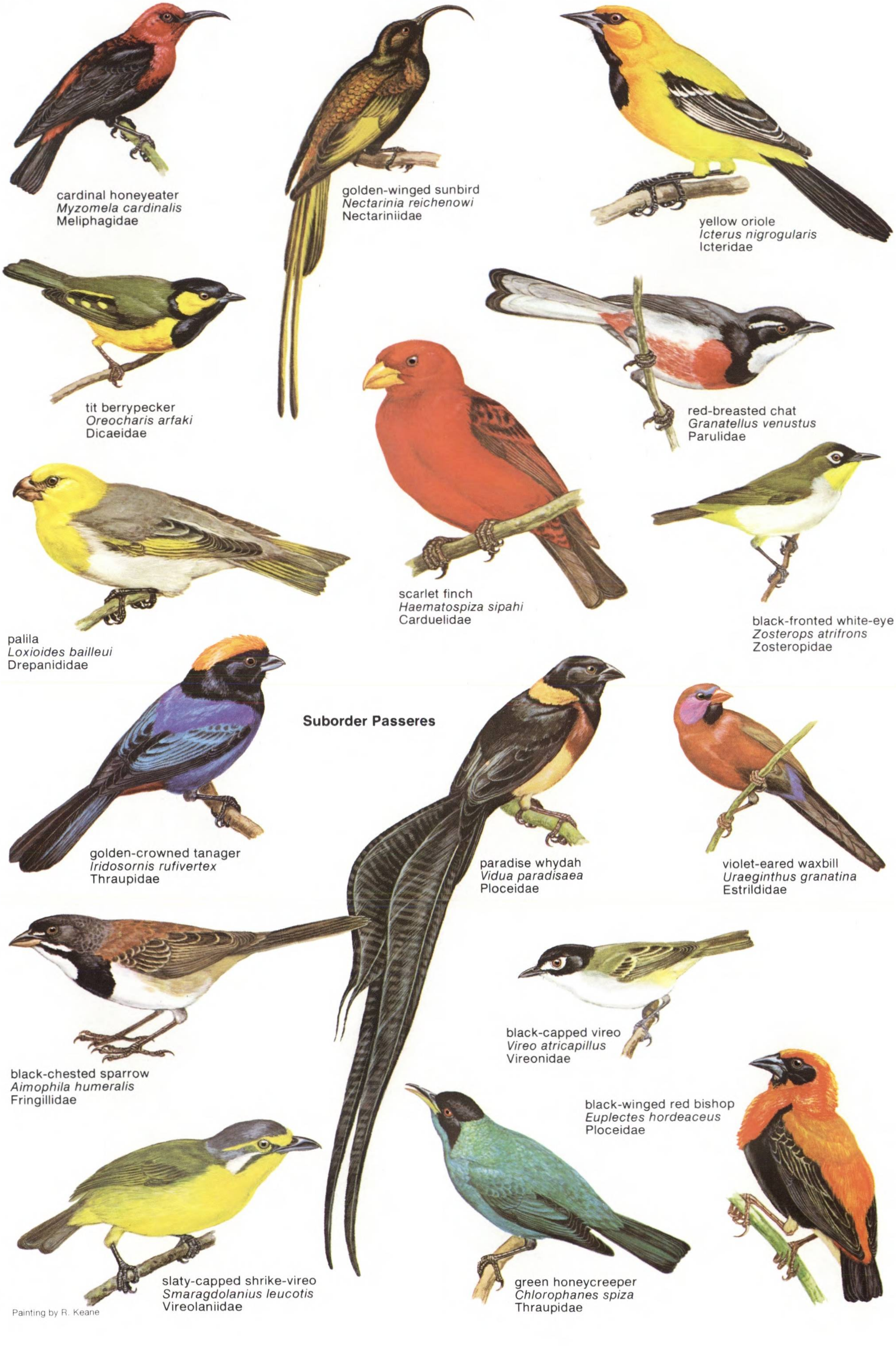

black-faced flycatcher
Monarcha frater
Muscicapidae


northern brubru
Nilaus afer
Laniidae


Fulleborn's longclaw
Macronyx fulleborni
Motacillidae


four-coloured bush shrike
Telophorus quadricolor
Laniidae

Suborder Passeres



cardinal honeyeater
Myzomela cardinalis
Meliphagidae

golden-winged sunbird
Nectarinia reichenowi
Nectariniidae

yellow oriole
Icterus nigrogularis
Icteridae

tit berrypecker
Oreocharis arfaki
Dicaeidae

red-breasted chat
Granatellus venustus
Parulidae

palila
Loxioides bailleui
Drepanididae

scarlet finch
Haematospiza sipahi
Carduelidae

black-fronted white-eye
Zosterops atrifrons
Zosteropidae

golden-crowned tanager
Iridosornis rufivertex
Thraupidae

paradise whydah
Vidua paradisaea
Ploceidae

violet-eared waxbill
Uraeginthus granatina
Estrildidae

black-chested sparrow
Aimophila humeralis
Fringillidae

black-capped vireo
Vireo atricapillus
Vireonidae

black-winged red bishop
Euplectes hordeaceus
Ploceidae

slaty-capped shrike-vireo
Smaragdolanius leucotis
Vireolaniidae

green honeycreeper
Chlorophanes spiza
Thraupidae

Painting by R. Keane

the fossil record is inadequate to elucidate the complete evolutionary history of the order, and the order has undergone an explosively radiative evolution that has resulted in a large number of anatomically similar species. Classification is further complicated by the persistence of intermediate and relict forms and a high degree of convergence (similarity due to different lines' being subject to the same types of natural selection) among distantly related groups. It is therefore necessary to examine each passerine species in careful detail (including its anatomy) to be certain of its affinities. With some 5,100 species in the order, this is a monumental task, one that is by no means complete. Traditionally, passerine families have been defined on the basis of a careful (usually anatomical) study of only a few individuals; the great majority of the birds have been assigned to established families solely because they externally resemble the studied species in plumage, bill, leg, and foot characters, as well as in what is known of their life history, behaviour, geographical distribution, etc. It is unfortunate that in the early years of avian taxonomy there was a strong tendency to try to include each newly discovered species, regardless of its geographical range, in some previously defined palearctic family.

Re-evaluation of early taxonomic characters

In the mid-20th century, taxonomists began re-examining the generally accepted 19th-century groupings of birds into families. In these investigations some of the features on which earlier classifications were based have been found to be convergent or too variable to be useful in certain groups (*e.g.,* bill shape, tarsal scutellation). Consequently, passerine taxonomists have been left with a rather sparse body of morphological information upon which to base a classification. In recent years ornithologists have made a concerted effort both to augment some of the century-old work on passerine anatomy and to explore new avenues of morphology, behaviour, reproductive patterns, biochemistry, and zoogeography to help define and relate the many families of perching birds. Much of this work is still in progress and has not yet been incorporated into classification systems. Among traditional and newly studied taxonomic features are external characters—such as rictal bristles and other specialized feathers, colours and patterns of the fleshy parts of the mouth, morphology of the bill and nostrils, colour patterns of adults and young; internal anatomical characters—such as the number of cervical (neck) vertebrae, the condition of the deep plantar tendons, anterior and posterior spines and processes of the sternum, syringeal muscles, palatal and other bones of the skull, feather tracts, jaw and tongue musculature, hyoid (tongue) apparatus, aortic-arch system, pneumatic fossa (cavity) of the humerus, and types of spermatozoa; biochemical analysis of substances—such as egg white, eye lens, plasma proteins, and hemoglobins; and an array of behavioral traits—such as reproductive behaviour, nest building, methods of scratching, etc. Thus, although the current system is by no means satisfactory, improvement is in the offing, and it is hoped that in the near future modern studies will lead to a more firmly drawn passerine classification.

Annotated classification. The classification and sequence of families given here is essentially that proposed in 1971 by American ornithologist Oliver L. Austin, Jr., based on a 1960 classification by Alexander Wetmore, with certain changes (see *Critical appraisal*). Some of the traditional diagnostic characters have been omitted because they have proved unreliable or because preliminary restudy indicates that the old generalizations do not hold true.

ORDER PASSERIFORMES (perching birds, or passerines)

Land birds with a characteristic "perching" foot; 4 toes (never webbed) joined at the same level, with the 1st toe (hallux) directed backward and never reversible. Oil gland unfeathered. Wing eutaxic (no gap between the 4th and 5th secondaries), usually with 9 or 10 primaries. The young altricial—*i.e.,* hatched almost or completely naked of feathers (a few exceptions), helpless, and requiring a considerable period of parental care. Groups marked with a dagger (†) are extinct and known only from fossils.

Suborder Eurylaimi

Syrinx simple, tracheobronchial, lacking intrinsic muscles and cartilaginous specialization; pessulus present. Sternum with anterior spina sternalis long and simple (exceptions) and with one pair of notches in posterior border. Hallux weak but about as long as the other toes. Clavicles well developed. Usually 15 cervical vertebrae (all other passerines have 14). Deep plantar tendons of different type than all other passerines.

Family Eurylaimidae (broadbills). Generally brightly coloured, chunky birds, with large heads, short necks; 12.5 to 28 cm (5 to 11 in.) in length. Bill broad and flattened, covered with a crest in some; gape wide. Front toes partially joined; 10 or 11 primaries. Other characters as in subordinal definition above; 14 species, primarily in forests or cloud forests. Indo-Malaya, Africa, Philippines.

Suborder Tyranni (majority of suboscines)

Syrinx usually more complex; muscles variable; pessulus present or absent. Sternum with short spina sternalis, forked (exceptions noted below); posterior border with one or two pairs of notches. Hallux strong. Clavicles well developed.

Superfamily Furnarioidea

Syrinx complex, tracheal; pessulus absent. Pterylosis (distribution of feathers) variable. Sternum with one or two pairs of notches.

Family Dendrocolaptidae (woodcreepers). Characteristically slender arboreal birds, olive brown to rufous, usually streaked or barred; 14.5 to 37 cm (5½ to 15 in.). Strong probing bills, laterally compressed, short and straight to long and downcurved; nares holorhinal (see Furnariidae). Short legs, powerful feet; front toes partially joined at base. Tail stiff with spiny tips. Posterior border of sternum with one pair of notches. Syrinx with cartilaginous specializations and two pairs of intrinsic muscles. Pterylosis with distinctive ventral tract type. Forty-eight species, forest or brushland. Mexico through South America.

Family Furnariidae (ovenbirds). A large and extraordinarily diverse group, an excellent example of adaptive radiation. In South America, especially in the southern Brazil–Argentina–Chile region, this family has evolved to fill a broad range of ecological niches; various furnariids look and behave like wrens, thrushes, dippers, creepers, larks, wood warblers, titmice, nuthatches, and even the nonpasserine woodpeckers (Picidae) and sandpipers (Scolopacidae). If such diversity can be summarized, ovenbirds are generally small, dull brown birds, darker above and paler below; 12 to 28 cm (about 5 to 11 in.). One distinctive character is the shape of the external nares (nostrils), which are schizorhinal (slitlike) at their posterior border rather than holorhinal (rounded), as in other passerines. Front toes partially joined at base. Sternum with one pair of notches (some exceptions). Syrinx with two pairs of intrinsic muscles. Pterylosis distinctive within suboscines but of same type as many oscines. Approximately 215 species, wide variety of habitats; Mexico through South America.

Family Formicariidae (antbirds). A large and diverse family, with loose-webbed plumage, generally in browns, grays, black, and white. Overall size range 9.5 to 37 cm (less than 4 to nearly 15 in.), including several large terrestrial species with long legs but very short tails (hence a deceptively short "body length"). Bill strong, hooked to variable degrees. Front toes slightly joined at base. Sternum with one or two pairs of notches, the two pairs found in most but not all of the long-legged terrestrial species. Syrinx of two types: the first with one pair of ventrally originating intrinsic muscles (largely the arboreal species), the second with no intrinsic muscles. Pterylosis variable; dorsal tract with reduced or absent posterior element (primarily the arboreal species); several different dorsal types among the terrestrial species. Wings generally short and rounded, flight weak; about 231 species, largely in dense forests and brushland. Neotropical; Mexico to northern Argentina.

Family Rhinocryptidae (tapaculos). A small, little-known group. Plumage dull, generally in grays, browns, and black; 9 to 25.5 cm (less than 4 to more than 10 in.). Bill distinguished by a movable flap (operculum) over the nostrils. Legs, feet, and claws strong. Wings short and rounded, flight feeble. Sternum with two pairs of notches. Syrinx with one pair dorsally originating intrinsic muscles (some exceptions). Pterylosis distinctive in ventral tract (exceptions). Twenty-seven species, grasslands, scrublands, and dense forest undergrowth. Central and South America, especially Chile and Patagonia.

Superfamily Tyrannoidea

Syrinx tracheobronchial; pessulus present or absent. Pterylosis: dorsal saddle with full-length apterium (exceptions). Sternum with one pair of posterior notches.

Family Cotingidae (cotingas). An extremely diverse, probably composite family; ranges from plain-coloured flycatcher-like to bright, extravagantly adorned birds: 9 to 45.5 cm (about 4 to 18 in.). Some with extraordinary plumes and lappets (the crow-sized umbrella bird, *Cephalopterus*); others renowned for their loud voices, fleshy wattles, and white plumage (white being unusual in land birds). The family as traditionally constituted can no longer be characterized anatomically: the syrinx is of two types; the tarsal envelope is of several different kinds; the joining of the front toes is variable; the dominant artery of the thigh is either the sciatic or femoral; the spina sternalis is forked except in some bell birds (*Procnias*). Several species with powder downs, a specialized type of feather otherwise known in passerines only in the oscine wood-swallows (Artamidae); 91 species, forests. United States–Mexico border to Bolivia, Peru, and Argentina.

Family Pipridae (manakins). Most piprids are fairly uniform in external appearance; generally small (8.5 to 16 cm [3½ to 6½ in.]), rather stubby, with short wings and tail (a few long-tailed species). Most males black with patches of brilliant colour (red, yellow, blue, etc.); females are generally drab olive green. Several species, such as the broad-billed manakin (*Sapayoa aenigma*), are externally and anatomically unlike the majority of manakins and may not be properly included in this family. Bill short, rather broad at base, notched, with a slight hook at tip. Some species with specialized feathers, particularly in the wings (also true of some cotingas and tyrannids). Third toe partially fused at base to 2nd or 4th. Syrinx highly variable, as family is presently constituted; 59 species, tropical and subtropical forest. Southern Mexico to Paraguay and northern Argentina.

Family Tyrannidae (New World, or tyrant, flycatchers). Large family of generally (but not exclusively) arboreal birds with plumage in grays, browns, olive greens, some with black, white, and yellow, occasionally brighter colours. Many with erectile crowns or crests, often more colourful than the rest of the plumage; length 7.5 to 40.5 cm (3 to 16 in.). Bill extremely variable but commonly broad, somewhat flattened, and hooked at tip; nostril rounded without an operculum or narrow with membranous operculum; most species with well-developed rictal bristles (stiff hairlike feathers around the mouth). Feet weak except in the few terrestrial species; front toes variably but never strongly fused. Syrinx characterized by 1 pair of intrinsic muscles (none or 2 in some), generally variable in other features. About 367 species, with a wide range of habitats. Northern Canada and Alaska through North and South America to Tierra del Fuego, also Falkland and Galápagos Islands. (The family common name of New World flycatchers serves to distinguish it from the large oscine family of Old World flycatchers, Muscicapidae.)

Family Oxyruncidae (sharpbill). A single, rare, little-known species, sometimes placed in the Tyrannidae. Length about 17 cm (7 in.); externally very much like flycatchers, with the same basic syringeal structure. Differences include a long, straight, and sharply pointed bill, surrounded at the base with short, fine, stiff feathers (not rictal bristles). Nostril narrow, covered by a broad horny operculum. Feet strong. Locally distributed in humid forest from Costa Rica to southeastern Brazil and Paraguay.

Family Phytotomidae (plantcutters). A small family of finch-like birds, gray or brown streaked with black, with areas of rusty red; about 17 cm (7 in.) long. Bill distinctive, stout and conical with a finely serrated edge. Legs short, feet large and strong. Syrinx simple, similar to that of some Cotingidae. Pterylosis: flank margin of ventral tract oblique, contrasting with truncated margin in other New World Tyrannoidea; 3 species, open brush and cultivated regions (where it is often considered a pest). Peru to Argentina and Chile.

Family Pittidae (pittas). A relatively homogeneous family; stout-bodied, long-legged, largely terrestrial, 15 to 28 cm (6 to 11 in.); with a wide range of colours in the loose-webbed plumage: reds, greens, blues, as well as browns, black, and white. Wings short and rounded but strong; tail very short. Syrinx simple, lacking intrinsic muscles and cartilaginous modifications; pessulus absent in most species. Egg-white protein pattern differs from that of all other suboscines. Twenty-three species, Old World tropics in forests and scrublands. Africa through Malaysia, Australia, Solomon Islands.

Family Xenicidae (New Zealand wrens). Small birds, 7.5 to 10 cm (3 to 4 in.), look and act much like true wrens (Troglodytidae). Considered evolutionary relicts, 2 species are strongly arboreal, the other 2 terrestrial. One of the latter, the now-extinct Stephen Island rock wren (*Xenicus lyalli*), was markedly terrestrial and may have been flightless; if so, it would have been the only known passerine to have completely lost the ability to fly. The 3 extant species are very weak flyers, with short wings and very short tails. Legs long and slender; 3rd and 4th toes fused basally, and all toes (especially the hallux) with long claws. Forest and scrub; New Zealand.

Family Philepittidae (asities and false sunbirds). The two species of asities (*Philepitta*) are black or yellowish green, rather pitta-like, with stout bodies and long legs. The two false sunbirds (*Neodrepanis*) are very different externally, looking much like true sunbirds (Nectariniidae), with which they were long classified; small, blue, yellow, and greenish with short tails and long, slender, curved bills. All philepittids have bare skin or wattles or both around the eyes. Currently classified in the Tyranni, but the simple structure of the syrinx and the unforked spina sternalis have led some authorities to suggest they are more closely allied to the Eurylaimi. Forests of Madagascar.

Suborder Menurae

Syrinx tracheobronchial, acromyodian, with 2 or 3 pairs of muscles inserted on the ends of the bronchial semi-rings; pessulus present. Sternum with short, forked spina sternalis, 1 pair of notches or none. Hallux strong. Clavicles not well developed.

Family Menuridae (lyrebirds). Among the largest members of the order in length, 75 to more than 100 cm (30 to 40 in.), with some weighing over 1,100 grams (about 2½ pounds), lyrebirds are aberrant passerines in virtually all features. In size, brownish plumage, and terrestrial habits, they look much like pheasants. The tail of 16 feathers is extraordinary: in males of *Menura superba* the outermost pair of feathers is curved like a lyre, the next 6 pairs are light and filamentous, and the innermost pair is modified into narrow "wires"; the tail of *M. alberti* is shorter and simpler; females of both species have rather long but not extremely modified tails. The alula (free digit on the leading edge of the wing) on the short rounded wings has 6 feathers rather than the usual 3 or 4. Rictal bristles present. Legs and feet strong and heavy. Sternum long, narrow, and thick, unique within passerines; posterior border may be entire or with 1 pair of shallow notches. Syrinx of the oscine type, but with only 3 pairs of intrinsic muscles; it is capable of a remarkable variety of loud ringing calls and the birds are famous for their vocal mimicry. Two species, dense forests; Tasmania (introduced) and eastern Australia.

Family Atrichornithidae (scrubbirds). Wholly unlike lyrebirds in external appearance, rather like large wrens, long-tailed, brownish, and terrestrial; 16.5 to 23 cm (6½ to 9 in.). One species, *Atrichornis clamosus,* long thought to be extinct but recently rediscovered. Both species virtually flightless, with very small wings and clavicles small and separated (not fused into a furcula, "wishbone"), a condition unique within passerines. Sternum unlike *Menura,* with 1 pair of deep posterior notches. Legs and feet very strong. Rictal bristles lacking. Similar to lyrebirds in the quality of their strong voices, nesting habits, and basic syringeal structure. Scrubbirds, however, have only 2 pairs of intrinsic muscles in the syrinx. Two species, scrublands; in western and eastern Australia.

Suborder Passeres (songbirds, or oscines)

Syrinx with 4 pairs of intrinsic muscles. Sternum with spina sternalis short and forked, and posterior border with 1 pair of notches. Hallux variable in strength. Clavicles well developed. All with same complex of syringeal muscles, with only minor variations.

Family Alaudidae (larks). A distinctive and well-defined group, the only oscines with back of tarsus rounded and scaled instead of sharp and unsegmented. Syrinx also unique among Passeres in lacking the usual bony pessulus. Small ground birds, 12 to 23 cm (about 5 to 9 in.), usually cryptically coloured in browns and grayish buffs, plain or streaked, lighter below (several species black below); sexes similar. Bill usually pointed, slightly downcurved; wings long, pointed; legs rather long, hind toe usually with long, straight claw. About 75 species inhabit open fields, plains, beaches of Old World, with only one species, the horned lark (*Eremophila alpestris*), also in North America south to Colombia. Many species migratory. Food: seeds, insects, and other invertebrates. Sing beautifully, often a soaring flight song.

†*Family Palaeospizidae (Palaeospiza).* A single fossil species known only from Oligocene of Colorado.

Family Hirundinidae (swallows, martins). A distinctive family placed low on the oscine family tree because of primitive syringeal characters, particularly the double bronchial rings, unique in Passeres. Small birds, 9.5 to 23 cm (about 4 to 9 in.), with compact plumage, often with metallic sheen, usually lighter below; sexes usually alike or nearly so. Wings long, pointed, primaries reduced to 9. Tail medium to long, truncate to forked. Legs short, feet small, weak. Bill short, flat, but wide gape enhanced by rictal bristles. Walk with difficulty but fly strongly, feed on insects caught in flight. Voice usually a twittering or squealing, sometimes melodious. The river martins of Africa and Thailand differ by large, brightly coloured bills, stronger feet, and syringeal characters. The 79 living species are found worldwide except polar regions and certain oceanic islands; all or nearly all migratory.

Family Dicruridae (drongos). Small to medium-sized birds, 18 to 63.5 cm (7 to 25 in.), the longest being those with exceptionally long tails; usually black with purple or greenish sheen, some crested or with spangled neck and head feathers, iris of eye usually red; sexes alike. Bill stout, arched, slightly hooked and notched; long, strong rictal bristles; legs short, feet stout. Wings long; tail variable, of 10 or 12 feathers, usually forked; some species with racquet tails. Arboreal birds that fly well but seldom long or far. Food mostly insects caught on the wing. Voice varied, melodious; capable mimics. The 20 species range through Africa, southern Asia, Malaysia, northern Australia, east to Solomons, in woodlands, savannas, cultivated regions.

Family Oriolidae (forest orioles, fig-birds). Medium-sized birds, 18 to 30.5 cm (7 to 12 in.); brightly coloured, predominantly in yellows, greens, and black; sexes unlike, female duller, young streaked below. Bill strong, pointed, slightly hooked; long, pointed wings with 10 primaries; medium to long tail of 12 feathers. Strictly arboreal birds that feed on insects, fruit; 28 species in forest and open woodland across Eurasia, in Africa, East Indies, Australia, Philippines. Flight strong, undulating. Voices loud, with harsh calls, but melodious songs. For New World orioles, see below Family Icteridae.

Family Corvidae (crows, jays, ravens, rooks, choughs, nutcrackers, and magpies). Medium to large passerines, including the heaviest oscines, the ravens (*Corvus*); 17.5 to 70 cm (7 to 28 in.). Ten primaries, the 10th (outer) always much shorter than 9th, but longer than primary coverts. Bill strong, powerful, longer than rest of head or nearly so; nostrils usually covered with bristles. Crows (*Corvus*) and allies, large, black or black and gray or white; wings long, tail shorter than wing. Jays generally smaller, often coloured in blues, greens, yellows; wings rounded; tail sometimes two-thirds of total length. About 102 species, almost cosmopolitan (absent from southernmost South America, Antarctica, some oceanic islands, introduced to New Zealand); varied habitats, prefer woodlands, open brushlands. Voices harsh, loud.

Family Callaeidae (wattlebirds). Medium-sized, 25.5 to 53.5 cm (10 to 21 in.); black, brown, or blue-gray, with fleshy blue or orange wattles at the gape; sexes may differ in size, wattles, and bill shape. Weak fliers but hop strongly. Sternum weak; 10-primaried wings short, rounded; tail long. Bill stout and short to long and curved; legs long, feet stout. Eat fruit, nectar, insects. A variety of musical notes and whistles. Three species limited to primeval forests of New Zealand; huia (*Heteralocha*) extinct; wattlebird (*Callaeas*) and saddleback (*Creadion*) rare.

Family Grallinidae (mudnest builders). Ten-primaried oscines of medium size, 19 to 50 cm (7½ to 20 in.). Differ from most corvids in lacking nasal bristles. Legs long, strong; wings long and pointed to short and rounded. Black and white or dark gray; sexes alike or unlike. Four species limited to Australia, western New Guinea, in woodlands, marshes, cultivated lands, usually near water. Weak fliers with peculiar jumping gait; all build deep, open-bowl nests of mud, stiffened with hair, feathers, grass. Eat insects, snails, seeds, soft fruit. Melodious whistles, harsh and plaintive notes.

Family Cracticidae (bell-magpies). Ten species of medium to large oscines, 25 to 58 cm (10 to 23 in.). Most black, white, and gray; some have brown phase; sexes alike or unlike. Large, heavy bill, slightly to strongly hooked; nostrils bare. Wings long, pointed; legs strong, medium to long. Australia, New Guinea, and nearby islands; gregarious inhabitants of brushy plains, open forests, mangrove shores, cleared lands. Eat insects, small animals, some fruit and seeds. Fine singers with ringing, gonglike calls.

Family Ptilonorhynchidae (bowerbirds). Ten-primaried oscines of medium size, 23 to 37 cm (about 9 to 15 in.); sexes alike in a few, but males usually much brighter coloured, often with nuchal (neck) crest but never with plumes or facial wattles. Stout bill, straight to slightly curved; wings rounded; legs and feet stout, hind toe shorter than middle toe. Seventeen species, in forests of New Guinea and northern Australia. Solitary, largely terrestrial birds; eat berries, seeds, fruits, insects, small animals. Males of most species build elaborate stages or bowers for display and courtship. Loud ringing calls, good mimics.

Family Paradisaeidae (birds of paradise). Small to medium-sized, 10-primaried, 14 to 117 cm (about 5½ to 46 in.), greatest length due to streaming tail feathers; greatly varied colours, most males with spectacular plumes on head, flanks, wings, or tail; some with wattles or bare skin on the head; females plain browns or grays. Bill slender to rather heavy, hooked or sickle-shaped in some. Wings rounded; legs short; feet rather stout. About 40 species, New Guinea, northern and eastern Australia, Moluccas, and adjacent islands. Solitary forest birds, rather weak fliers; males of most species have elaborate courtship displays. Eat fruit, seeds, insects, small animals. Prolonged whistles, loud shrill calls.

Family Paridae (titmice and chickadees). Ten-primaried oscines with short to medium, rounded wings; tail short to long; bill rather stout, pointed, shorter than remainder of head; nonoperculate nostrils concealed by thick feathers. Small birds, 7.5 to 20 cm (3 to 8 in.), with thick plumage usually strongly patterned in grays, yellows, brown, black, or white, never streaked, barred, or spotted; sexes usually alike. About 64 species in forests and brushlands of the Philippines, Malaysia, Eurasia, Africa, and North America to Guatemala. Active, gregarious little birds; feed on insects, seeds. Chattering notes, whistled calls.

Family Certhiidae (creepers). Small, slender, climbing birds, 9.5 to 19 cm (3½ to 7½ in.), with curved bills as long as or longer than rest of head; operculate nostrils free of bristles or feathers; rictal bristles absent. Legs short, thin, outer toe always shorter than middle toe but much longer than inner toe; claws long, hind claw as long as or longer than hind toe; wings rounded or pointed; tail feathers long, stiff, with pointed tips. Brownish above, streaked with black or light brown, white below; sexes alike. Six species (with the inclusion of the African spotted creeper, *Salpornis,* a controversial point) in temperate woodlands of Eurasia, Africa, North America, south to Nicaragua. Solitary birds, spiral up tree trunks from base probing bark for insects; two species inhabit rocky cliffs. Soft calls, sweet but weak songs.

Family Sittidae (nuthatches). Small, stocky, climbing birds, 9.5 to 19 cm (3½ to 7½ in.), with thin, pointed, usually straight bills; rounded, nonoperculate nostrils partly concealed by feathers; short rictal bristles present; tarsus short; long, laterally compressed claws on large toes, hallux equal to outer toe, inner toe reaching only to second joint of middle toe; wings rather long, pointed; tail short, square, soft. Typically gray to blue above, white or brownish below; sexes alike or nearly so. About 22 species; typical nuthatches (*Sitta*) distributed across North America and Eurasia to Malaysia and the Philippines, aberrant forms in Philippines, Australo-Papuan region. Typically in forests, a few in rocky areas. Forage on tree trunks and large branches for insects, eat some seeds and small fruits. Simple call notes and songs.

Family Climacteridae (Australian treecreepers). Six species of small, creeper-like climbing birds, 12.5 to 17.5 cm (5 to 7 in.); of uncertain ancestry and affinities. Legs short; toes long, claws long, curved, strong, especially that of hallux; tail rounded, soft; bill long, somewhat downcurved. Grayish brown to black above, streaked below, often a lighter eye stripe and wing bar; sexes similar but with slight recognizable differences. Forests of Australia and New Guinea; rather solitary, sedentary birds; feed by spiralling up tree trunks, some forage for insects on ground. Voice high-pitched whistles.

Family Panuridae (bearded tits, parrotbills). Small, titmouse-like birds, 10 to 17.5 cm (4 to 7 in.), distinguished (with one exception, the bearded tit, *Panurus biarmicus*) from all other oscines by the strongly compressed bill, much shorter than head, markedly curved convexly, both tomia (cutting edges) sinuated; nostrils concealed by feathers; wings short, rounded; tail rather long, graduated. Plumage soft, fluffy, plain brownish above, lighter below; sexes alike or nearly so. About 19 species in brushy grasslands, scrublands, and thickets of temperate Eurasia; active, gregarious, travel in small flocks searching for seeds, insects, berries. Low twittering call notes.

Family Chamaeidae (wrentit). Small bird about 16.5 cm (6½ in.) long, with thick fluffy plumage; plain brown above, lighter and streaked below. Bill short, rounded, pointed, slightly downcurved; nostrils exposed, operculate; rictal bristles distinct; wings short, rounded; tail long, graduated; sexes alike. The single species ranges in brushlands and forest edges from Oregon to northern Baja California; usually in pairs. Eats insects, small berries. Voice: a loud whistle on one pitch. Often classified in the Timaliidae.

Family Timaliidae (babblers). A poorly defined, diverse group of more than 250 species, with uncertain affinities and lineage. Small to medium-large, 9 to more than 40 cm (3½ to 16 in.). Plumage soft, fluffy; bill highly varied, small and weak to long, straight, downcurved, often hooked. Wings short, rounded, fit close to body; legs rather large, strong. Mostly dull coloured, a few brightly coloured; sexes alike and unlike. In woodlands, scrub, and brushlands of eastern and central Eurasia, Africa, Madagascar, Philippines, Malaysia, and Australia; solitary or gregarious, usually arboreal, some terrestrial; all insectivorous, some also eat fruit, berries. Many species with noisy, harsh calls, musical songs.

Family Campephagidae (cuckoo-shrikes and minivets). About 70 species of small- to medium-sized, forest-living birds, 12.5 to 35.5 cm (5 to 14 in.). Africa, India to Japan, East Indies, Australia. Rather stout, slightly downcurved bill, notched, hooked at tip; nostrils partly concealed by short bristles; wings medium, pointed; long tail rounded or graduated; legs short, feet weak to strong. Feathers of back and rump usually with heavy shafts, thickly matted and loosely attached. Cuckoo-shrikes plain gray, black, or whitish birds, often with barred underparts; minivets brightly coloured in reds, yellows. Sexes often completely dissimilar. Eat insects, berries, small fruits. Noisy birds, with whistles and harsh calls.

Family Pycnonotidae (bulbuls). Medium-sized birds, 14 to 28 cm (5½ to 11 in.), with soft, fluffy plumage, especially soft on lower back and rump; many species with hairlike, vaneless feathers on nape of neck. Bill usually slender, slightly downcurved; rictal bristles well developed; feet and legs rather small; wings short; tail medium to long. Usually drably coloured in gray, brown, greenish, some with yellow, white, or red patches on head and under tail; sexes alike. About 119 species in Africa and across southern Asia to Japan, Philippines, Moluccas, Borneo; in woodlands, brushlands, cultivated regions. Eat berries, fruit, some insects. Noisy, some good singers, mimics.

†*Family Palaeoscinidae (Palaeoscinis).* A fossil species known only from the Miocene of California.

Family Irenidae (leafbirds, ioras, fairy bluebirds). Small to medium birds, 13 to 25 cm (5 to 10 in.); bill fairly long, slightly downcurved or hooked; legs short; wings rounded; tail square to rounded. Like bulbuls, some have hairlike feathers on nape and long fluffy rump feathers but are more brightly coloured, brown or black with contrasting yellow, green, or blue; sexes usually dissimilar. About 14 species in forests and cultivated lands from India to Philippines, south to Borneo, Java, Sumatra. Eat fruit, berries, buds, some insects. Good singers, musical whistles and fluty notes.

Family Cinclidae (dippers). The only largely aquatic oscines; small, 14 to 19 cm (5½ to 7½ in.), birds with plump bodies, short concave wings, short, square or rounded tail. Tarsi and toes long and stout. Bill slender, straight, pointed; nostrils with broad operculum; rictal bristles absent. Dense compact plumage plain gray, brown, or black, some with white patches on throat, breast, head; sexes alike. Four species, found along swift, rocky streams in Eurasia, North Africa, and North and South America, south to Argentina. Food: insects, aquatic larvae, and other organisms gleaned from stream bottoms. Have shrill penetrating whistled calls and chattering songs.

Family Troglodytidae (wrens). Small, 9.5 to 22 cm (3½ to 8½ in.), chunky birds, mostly brown, usually barred, spotted, or streaked, with white, black, or browns; sexes alike. Bill slender, medium to long, often downcurved; nostrils with operculum; rictal bristles usually indistinct but sometimes obvious with 1 or 2 well developed. Stout legs and feet, front toes partly joined at base. Wings short, rounded, well-developed 10th (outermost) primary, at least half as long as 9th; tail short, square to rounded, often carried cocked up. The 59 species range through most of North and South America; 1 also found in Eurasia and North Africa; inhabit brushlands, forests, forest edges, rocky slopes, deserts, grassy marshes. Most forage on or near ground in undergrowth for insects, worms, other invertebrates. Highly developed song, musical, variety of bubbling, flutelike to growling notes, harsh chattering calls; some species duet.

Family Mimidae (mockingbirds, catbirds, thrashers). Slender, medium-sized, 20 to about 30 cm (8 to 12 in.); bill medium to long, nearly straight to strongly downcurved; nostrils always exposed, with overhanging membrane; rictal bristles present, but few and somewhat weak. Legs rather long, feet strong, middle toe joined at base to outer toe but not to inner toe. Wings short, rounded; tail long. Coloured brown, gray, black, bluish, underparts usually pale, often white, spotted, or streaked, some solidly coloured, some with white in wings or tail; sexes alike. Exclusively New World, the 31 species range from southern Canada, the West Indies, Galápagos, to southern Argentina and Chile. Eat fruit, seeds, insects. Famous as fine singers and mimics.

Family Turdidae (thrushes, bluebirds, nightingales, wheatears, robins, chats). A large, almost cosmopolitan group of about 305 species of small- to medium-sized birds, 11.5 to 33 cm (4½ to 13 in.). Bill rather slender; legs and feet fairly stout, tarsus usually booted (smooth sheath, not divided into scales); 10-primaried wing rounded to pointed; tail medium, truncate or graduated, forked in a few. Colours brown, blue, gray, often blended, or in bold black, white, yellow, or red patterns; sexes alike or unlike; young usually spotted below. Virtually worldwide; absent originally only from New Zealand (now introduced there), some oceanic islands, Antarctica, and parts of the Arctic. Usually arboreal, many terrestrial, in varied habitats—forests, deserts, brushlands, grasslands, cultivated fields. Eat mostly fruit, insects, also seeds, leaves, worms, and mollusks. Most are fine singers.

Family Sylviidae (Old World warblers, kinglets). Typically small, 9 to 26 cm (3½ to 10 in.), with slender bill; longitudinal nostrils with an operculum; medium rounded wings of 10 primaries; short to medium legs. Plumage usually browns, grays, olive greens with little pattern; some streaked or barred; some brightly, boldly marked; sexes alike or similar, young never spotted below. Nearly 300 species, worldwide except in polar regions, South America, some oceanic islands, mostly in woodlands, some in brushlands, marshes. Food essentially insects; voices pleasant, varied, well-developed song in some.

Family Polioptilidae (gnatcatchers, gnatwrens). Dainty, slender, tiny, 10 to 14 cm (4 to 5½ in.), with long, thin, pointed bills, operculate nostrils partly exposed, and rictal bristles. Rounded wing with 10th primary much less than half as long as 9th; long, rounded tail constantly moving. Blue-gray or brown above, lighter below, most with white in outer tail feathers, some with black markings on head; sexes alike or nearly so. The 11 species are active arboreal birds of open forests and semi-arid deserts from southern Canada to Argentina. Food essentially small insects. Rather weak songs of trills and warbles, simple short call notes.

Family Pachycephalidae (whistlers, shrike-thrushes, thickheads). Small- to medium-sized, stout-bodied, 13 to 28 cm (5 to 11 in.); roundish heads, rather heavy bill sometimes hooked at the tip. Wings rather long, pointed, with very short 10th primary; tail medium long, usually rounded. A few species crested or with wattles or bare patches at base of bill. Most are greenish gray to brown above, lighter below, many with yellow or dull red markings; sexes alike, a few unlike; many juvenile plumages spotted or like that of female. About 42 species in Australo-Papuan region, Malaya, Philippines, Oceania; in forests, brushlands, mangroves, savannas. Diet predominantly insects with some fruit. Melodious fluty calls, pairs often duet.

Family Maluridae (wren-warblers, emu-wrens). Small-bodied birds, 7.5 to 25 cm (3 to 10 in.), that carry the long tail cocked up over the back. Bill small, weak; wings short, rounded; legs and feet medium. Emu-wrens (*Stipiturus*) have rectrices reduced to 6 loose-barbed shafts. Most species are brightly patterned in contrasting browns, reds, shiny blues, black, and white; sexes alike or unlike. About 83 species in Australo-Papuan region, including New Zealand; in forests, scrublands, heaths. Food largely insects and other invertebrates. Many are good singers, mimics.

Family Muscicapidae (Old World flycatchers). A large (about 340 species) family of small insectivores, 7.5 to 22.5 cm (3 to 9 in.); the paradise flycatchers (*Terpsiphone*) with long tails, to 53.5 cm (21 in.). Typically with flat, broad bills, well-developed rictal bristles; short, weak legs and feet. Wings short and rounded to long and pointed; tail short and narrow to long and fanned or with long central plumes. Varied colours, many plain browns, grays, some with bright blues, reds, and black and white; some crested or with facial wattles; sexes alike and unlike, young usually spotted. Eurasia from tree line south through Africa, Australia, and in Pacific Islands to Hawaii; forests, scrublands, cultivated and riverine areas. Voices usually weak, monotonous, but well-developed, song in a few.

Family Prunellidae (accentors, hedge sparrows). Small, drab, 12.5 to 17.5 cm (5 to 7 in.); slender, pointed bills, wide at base, culmen (ridge of upper bill) slightly rounded. Tarsus rather short, feet strong. Wings rounded to pointed with very short 10th primary; tail shorter than wing, square or emarginate. Colour browns or grays, usually streaked or spotted; sexes similar. Twelve species across Eurasia and northern Africa; various habitats, parks and gardens, brushlands, barren mountain slopes below snow line. Feed on or near ground on insects, berries, seeds. Thin chattering, twittering songs, metallic calls, often in flight.

Family Motacillidae (pipits, wagtails). Small, slender-bodied ground birds, 12.5 to 23 cm (5 to 9 in.). Pipits similar to larks in appearance but differ in having a bilaminate tarsus and pointed wing with 9 primaries. Wagtails have longer tails, brighter colours. Bill thin, pointed; legs long, slim, with elongated hind toe and hind claw (with some exceptions); tail usually edged with white or yellow. Plumage brown, streaked, mottled, sexes similar (pipits); or marked in bold black, white, or yellow patterns, sexes unalike (wagtails). About 53 species, worldwide except polar regions and some oceanic islands; in open grasslands, deserts, shores, cultivated areas. Feed on insects, spiders, mollusks, some vegetable matter. Simple repetitive song often given in flight, and short, sharp call notes.

Family Bombycillidae (waxwings). Small, arboreal, soft-plumaged, crested, 15.5 to 19 cm (6 to 7½ in.); distinguished by waxy red "droplets" at tips of secondary wing feathers of most individuals. Bill short, swollen, slightly hooked; nostrils almost concealed by feathers, rictal bristles absent. Long, pointed wing with rudimentary 10th primary; square to slightly rounded, short tail. Velevety plumage blended with browns, grays, and yellow; tail tipped with yellow or red; sexes alike. Three species in temperate Northern Hemisphere, in evergreen or birch forests. Eat berries, fruit, buds, flowers, insects. Weak chattering song, soft lisping calls.

Family Ptilogonatidae (silky flycatchers). Small, arboreal, crested, 18 to 24.5 cm (7 to 9½ in.). bill short, broad, deeply cleft; nostrils exposed, bordered by membrane, rictal bristles present. Wings and legs rather short; 10th primary well developed; tail long. Soft silky plumage of solid grays, black, brown, white spots in wings and tail, some with yellow markings. Four species, found from southwestern U.S. to Panama; in brush country, desert scrub, open forest. Rather shy, active birds that feed on berries and on insects caught in flight. Weak warbling song.

Family Dulidae (palm chat). Medium-sized, about 19 cm (7½ in.); deep laterally compressed bill, strongly curved culmen, circular nostrils wholly exposed. Wings rounded, 10th primary less than half as long as 9th; tail longish; legs and toes stout. Plumage stiff and harsh, olive brown above, yellowish-white streaked with brown below; sexes alike. The single species limited to cultivated areas and open woodlands of Hispaniola and Gonave Island, West Indies. Lives on fruits and flowers. Noisy, gregarious birds with harsh chattering notes.

Family Hypocoliidae (hypocolius). Medium-sized, 17.5 cm (7 in.); short, broad bill, moderately curved culmen, operculate nostrils; nasal bristles lacking and rictal bristles poorly developed. Tarsus rather short, heavily scutellated (scaled). Soft plumage blue-gray above, lighter below, with black facial, wing, and tail markings, white wing tips, slightly crested; sexes alike. The single species inhabits semi-arid scrublands, palm groves, gardens in Tigris-Euphrates valley of Iraq, wanders to Afghanistan, western India, Arabia, Red Sea coast. Fruit eater with low calls and weak song.

Family Artamidae (wood-swallows or swallow-shrikes). Chunky-bodied, medium-sized, 15 to 21 cm (6 to 8½ in.); unique among oscines in having powder downs. Bill stout; broad at base, moderately long, decurved, pointed; legs short; feet strong. Wings long, pointed; tail short, nearly square. Plumage compact, soft, in plain solid browns, grays, or black above, usually lighter below; sexes similar. Ten species in open lands, forest clearings, Australia east to Fiji Islands, north to Philippines, Indochina, India. Live on insects caught in flight. Voice a harsh nasal twittering.

Family Vangidae (vanga shrikes). Small- to medium-sized, 12.5 to 32.5 (5 to 13 in.), 10-primaried, arboreal oscines of varied aspect; some resemble flycatchers, thrushes, nuthatches, or typical shrikes. Bill stout, heavy, hooked, notched; with enlarged ridge in one, thin, downcurved in another; legs and feet strong; wings fairly long, rounded; tail square or rounded, moderately long. Plumage rather soft, loose, typically black or blue above, white below, some with brown or gray patches; sexes alike or unlike. Thirteen species, restricted to forests, scrublands, mangrove swamps, of Madagascar. Eat insects, small reptiles, amphibians. Chattering calls, harsh notes, shrill whistles.

Family Laniidae (shrikes). Rather small- to medium-sized, 15 to 36 cm (6 to 14 in.); 10-primaried, with proportionately large heads; stout, strong, sometimes toothed, sharply hooked bills; strong legs with sharp claws; tarsus scutellate anteriorly, lamellate (plated) laterally; wings medium; tail usually long, narrow. Plumage soft, black, gray, or brown above, usually paler below in true shrikes; yellows, reds, greens in some African bush shrikes; sexes alike or unlike. Essentially Old World, 64 species across temperate Eurasia, Africa, east to Philippines, and south to New Guinea and Timor; 2 species in North America to southern Mexico; usually solitary birds, in open forests, clearings, brushlands, cultivated areas. Food large insects, small vertebrates; some impale prey on thorns for storage and to tear apart. Voices varied, some with well-developed song, several species sing antiphonally.

Family Prionopidae (helmet shrikes). Medium-sized, 19 to 25.5 cm (7½ to 10 in.); distinguished from preceding family by their fully scutellate tarsus and stiff forehead feathers projecting forward, almost concealing the nostrils; most with conspicuous wattles surrounding the eyes. Bill stout, hooked; legs short, strong; wings and tail medium to long. Plumage black above, white or buffy below in bold patterns; sexes similar. Nine species, in brushlands, scrub, and forests of Africa, south of the Sahara. Gregarious arboreal birds that hunt in small flocks for insects. Harsh chattering notes, nasal humming sounds, snap bills audibly.

Family Sturnidae (starlings, mynas, oxpeckers). Stocky, medium-sized, 16.5 to 42 cm (6½ to 16½ in.); very short but visible 10th primary. Bill pointed, straight, or slightly arched, swollen near the tip in oxpeckers (*Buphagus*); tongue flat, not tubular (as in some relatives). Legs and feet stout, strong, tarsus with unbroken plates behind. Typically dark coloured with metallic sheen, some brown or gray with white, yellow, or red markings, some with facial wattles, a few crested; sexes alike or unlike. About 107 living, 4 recently extinct species in all types of wooded and agricultural lands; temperate Eurasia through Africa, northern Australia, East Indies, east to Tuamotu Archipelago; a few introduced widely elsewhere (*e.g.*, North America). Almost omnivorous; wide range of vegetable and animal foods. Garrulous, varied notes, calls, whistles, some excellent mimics.

Family Meliphagidae (honeyeaters). Small- to medium-sized, 10 to 40 cm (4 to 16 in.); long, protractile, brush-tipped tongue curled at the sides to form a tube. Bill slender, pointed, downcurved, upper cutting edge serrated; nostrils unfeathered, with leathery operculum. Wings long, pointed, 10th primary about half the length of 9th; tail medium to long. Legs short to medium, tarsus scaled anteriorly. Drab browns, yellows, grays, or bold patterns of black, red, white; sexes alike or unlike. About 168 living, 4 recently extinct species, from Australia and New Zealand through the Papuan region, north to Marianas, east to Hawaii, 2 species in southern Africa. Habitat forests, brushlands, cultivated lands. Food chiefly nectar, insects, some fruit. Voices loud, varied; musical song in many.

Family Nectariniidae (sunbirds, spider hunters). Small, 9 to 22 cm (3½ to 8½ in.). Long, slender, downcurved, pointed bill, finely serrated near tip; tongue not brush-tipped but partly tubular, projectile, divided at tip; nostrils rounded, open, operculate; no rictal or nasal bristles. Wings rounded, short, 10th primary variable in length but always present; tail square, medium to long, pointed, sometimes with elongated central feathers. Legs short, stout, tarsus with anterior transverse scales; hallux and claws short. Many sunbirds with vivid colours, metallic sheens, and bright patches in male; females much duller; other sunbirds and spider hunters dully coloured in both sexes. About 116 species, in forests, scrub country, mangroves; Asia and Africa south of Sahara, east to central China, Philippines, Solomons, south through Malaya to northern Australia. Active arboreal birds, sunbirds are Old World counterparts of New World hummingbirds but not as skilled fliers. Feed on nectar, insects, some small fruit. Weak call notes, faint songs.

Family Dicaeidae (flower-peckers, pardalotes, diamond birds). Small chunky birds, 7.5 to 18 cm (3 to 7 in.), with short necks, legs, tails. Bill usually short, stout, relatively straight; slender and curved in a few, edges of distal (outer) 3rd serrated; distal half of short tongue deeply cleft, the edges curled into 2 slender semitubular tips. Wings rather long, 10th primary usually vestigial, absent in a few. Bright-coloured males dark, glossy (rarely metallic) above, lighter below, often with red or yellow patches on crown, rump, breast; females usually much duller; some species dull-coloured in both sexes. About 58 species in forests, scrublands; India to southern China, Philippines, east through Solomons, and south to Australia. Active arboreal birds, feed around flowers on nectar, small insects, berries (particularly mistletoe). Sharp metallic twittering calls, a few have warbling song.

Family Zosteropidae (white-eyes). Relatively uniform group of little birds, 10 to 15 cm (4 to 6 in.) long, yellowish or reddish-brown to olive green above, lighter below, typically with a conspicuous white eye-ring; sexes alike. Bill short, slender, pointed, slightly decurved; tongue brush-tipped; rictal and nasal bristles absent. Wings short, rounded, 10th primary usually lacking; tail medium, square; legs and claws short, feet strong. About 90 species in forests, brush country, mangroves; Africa south of Sahara, across southern Asia to Korea, Japan, south to Australia, New Zealand, east to Carolines, Samoa. Travel in restless flocks, feed on nectar, insects, fruit. Weak voices, pleasant twittering, warbling songs.

Family Cyclarhidae (pepper-shrikes). Rather heavily built, medium-sized, 14.5 to 17.5 cm (5½ to 7 in.), with large heads; stout, strongly hooked, laterally compressed bills; nostrils rounded, exposed; rictal bristles present but weak. Wings short, rounded, 10th primary half length of 9th; tail medium; legs and feet strong. Loose-webbed plumage olive green above, yellow to buffy-white or gray below, rufous stripe over eye; sexes similar. One species southern Mexico to Guianas, Uruguay, northern Argentina; a second in Colombia and Ecuador; open forest, brushland. Weak fliers, travel singly or in pairs, moving deliberately through foliage. Eat large insects, fruit. Repetitive warbling song, harsh scolding calls.

Family Vireolaniidae (shrike-vireos). Small, poorly known; 14.5 to 18.5 cm (5½ to 7 in.). Rather stout, hooked bill; nostrils oval, exposed; rictal bristles inconspicuous. Wings short, rounded, 10th primary reduced to half length of 9th; tail short to medium; legs short, stout. Plumage silky, loose-webbed, greenish above, yellowish to white below, head marked with yellow, gray, blue, black, or white; sexes similar. Three species; in rain-forest treetops from southern Mexico to central Brazil. Food fruits, insects. Loud repetitive whistled song.

Family Vireonidae (vireos, greenlets). Rather plainly coloured, small arboreal birds, 10 to 17.5 cm (4 to 7 in.), mostly brownish-gray to olive green above, yellow, grayish, or white below; plumage never streaked or spotted; some with light eye-rings, eye stripes, wing bars; sexes alike. Bill fairly heavy, slightly hooked and notched; nostrils ovate, operculate, partly exposed; rictal bristles inconspicuous; legs short, strong. Wings long, pointed to short, rounded, 10th primary very short or vestigial. About 37 species in all types of woodlands from central Canada to Uruguay and Argentina. Usually solitary inhabitants of forest edges; seek insects on leaves and branches, eat some berries, fruit. Compulsive singers; songs of repeated, often melodious, phrases; harsh scolding notes.

Family Drepanididae (Hawaiian honeycreepers). Small forest birds, 11.5 to 22 cm (4½ to 8½ in.), limited to the Hawaiian Islands; 14 species living, 10 of them rare and local; 8 species are recently extinct. Bill extremely varied among species: short to long, thin to thick, straight to extremely downcurved, pointed to hooked, but never serrate or notched. Wings pointed, 10th primary vestigial or absent; tail medium, truncate or slightly forked; legs short to medium; feet strong. Plumage plain brown, olive green, yellow, red, gray, or black, never metallic or glossy; sexes unlike or similar with female smaller. Found solitary or in small flocks. Feed on nectar, insects, seeds, fruit. Warbling songs, clear calls.

Family Parulidae (wood warblers, bananaquits). Dainty, small, 10 to 18.5 cm (4 to 7½ in.); pointed wings of 9 primaries, medium 12-feathered tail. Bill usually slender, pointed, culmen slightly downcurved, flattened with pronounced rictal bristles in a few but never notched or hooked; tongue moderately slender, tip variably bifid (divided) or fimbriate (fringed). Legs medium, hind claw never elongated. Colours varied but never metallic or glossy; grays and olive browns dominant ground colours, often brightly patterned with yellows, reds, blue, black, or white; sexes alike or unlike. About 119 species in forests and brushlands from tree line of North America to southern South America. Active, arboreal (a few terrestrial); feed mostly on insects, occasional fruits, berries, seeds, 1 group (bananaquits) sip nectar. Voices typically weak, thin, high-pitched; a few with loud and well-developed songs.

Family Zeledoniidae (wrenthrush). The single species is small, 11.5 cm (4½ in.), with 9 primaries. Bill weak and flattened vertically; legs long, feet large, strong; wings and tail short, rounded. Plumage soft, olive brown above, slate gray below, crown black-bordered orange; sexes are alike. Resident in humid cloud forests of Costa Rican and western Panamanian highlands. Shy, retiring birds with weak, fluttery flight, seldom fly more than 20 m (65 feet), preferring to creep and hop on forest floor. Food consists mainly of insects. Voice thin, high-pitched, repetitive whistles.

Family Icteridae (orioles, troupials, oropendolas, caciques, blackbirds, grackles, cowbirds, meadowlarks). A heterogeneous group, medium to large, 16 to 54 cm (about 6 to 21 in.). Straight, pointed, unnotched, conical bill, with culmen extending more or less onto forehead; operculate nostrils, never concealed; rictal bristles lacking. Wings mostly long, pointed; tail short to rather long; legs and feet strong. Plumage often solid black with metallic gloss, or multicoloured in bold patterns of black and red, orange, yellow, brown; sexes usually differ, female smaller. About 88 species, throughout Western Hemisphere (except polar regions), in a wide variety of habitats. Many species gregarious; some colonial or semicolonial nesters. Eat a wide range of animal and vegetable foods. Usually loud-voiced, with harsh, whistling, or bubbling calls; well-developed song in some.

Family Tersinidae (swallow-tanager). Single tanager-like species, 15 cm (6 in.), differing from Thraupidae mainly in peculiar palate, expandable throat, and broad, flat, swallowlike bill; longer wings, shorter legs. Male shining turquoise with black throat, face, sides barred black, white below; female green with grayish-white barring. Lowland and mountain forests; Panama, Trinidad, south into Argentina. Somewhat gregarious; strong fliers; feed on insects, fruit. Voice monotonous chirps, almost no song.

Family Thraupidae (tanagers, euphonias, honeycreepers, and diglossas). Small to medium, compactly built, 9 primaries; 8 to 30 cm (about 3 to 12 in.), most under 20 cm. Bill with commissure (meeting of edges) not abruptly angled or bent at the base; mandibular tomium not distinctly angled (never toothed) near the base. Bill usually conical, short to medium, generally notched, toothed, or hooked at tip; rictal bristles present. Nectar-feeding species with specialized bill and brush-tipped tongue. Wings pointed, short to long; tail short to medium, truncate, emarginate, or rounded. Generally brightly coloured, in contrasting patterns of black, white, yellow, red, blue, green, brown; sexes alike or unlike. About 223 species in forests, brushlands; temperate North America to Brazil, Argentina. Arboreal birds; flight strong but not sustained in non-migratory species. Eat berries, fruit, insects, nectar. Voices varied, song well developed in a few but usually short calls and warblings.

Family Catamblyrhynchidae (plush-capped finch). Finchlike, 9 primaries, 15 cm (6 in.), with stubby, conical, slightly hooked bill and a distinctive patch of short, stiff, velvety tipped, orange-yellow feathers on the forecrown. Wings short, rounded; tail medium; legs stout, feet strong. Male bluish-gray above, chestnut below, nape and sides of head black; female similar but duller. Mountain forests of Colombia, Venezuela south to Peru, Bolivia. A little-known species of uncertain relationships and systematic position.

Family Fringillidae (New World seedeaters). Small, 9 primaries; 10 to 27 cm (4 to 11 in.), with short, stout, conical, pointed bills. Commissure distinctly angled or deflexed at the base, mandibular tomium elevated, often toothed, gonys (midline ridge of lower bill) more than half length of upper bill; rictal bristles usually obvious. Wings short, rounded to long, pointed; tail short to long, tarsus relatively long. Colours and patterns varied; sexes alike or unlike. About 315 species almost cosmopolitan (absent from Oceania, East Indies, Madagascar, Antarctica; introduced in Australia, New Zealand) in all terrestrial habitats. Gregarious or solitary, terrestrial or arboreal birds; food mainly seeds, some other vegetable matter, insects. Call a simple chirp, often well-developed song; flight song rare. Family includes buntings, juncos, towhees, grosbeaks, cardinals, longspurs, Darwin's finches, chaffinch, yellow hammer, most sparrows.

Family Carduelidae (goldfinches, siskins, rosefinches, redpolls, crossbills, bullfinches, hawfinches, canaries). Small to medium-sized, 10 to 25 cm (4 to 10 in.), with short, stout to slender pointed bills (mandibles crossed in one genus); gonys less than half length of upper bill. Wings somewhat rounded, medium length, 10th primary present but less than half length of 9th; tarsus relatively short. Colours variable but browns, reds, and yellows predominate; some streaked or mottled, some plain patterns; sexes alike or unlike. About 112 species in woodlands, brushlands, worldwide except Pacific Islands; introduced in Australia. Gregarious birds with strong, undulating flight. Eat seeds, buds, berries, some insects. Highly developed songs, sing in flight.

Family Estrildidae (waxbills, some finches). Small, 7.5 to 15 cm (3 to 6 in.), bill stout, short, pointed; gonys less than half the length of upper bill. Wings rounded, short to pointed, medium length; 10th primary present but almost vestigial; tarsus relatively short. Great variety of bright and somber colours and patterns; sexes alike or unlike. About 107 species; forest edges, clearings, grasslands, and marshes of Africa, southern Asia, East Indies, Australia, some South Pacific islands. Highly gregarious, active, ground feeders. Food: seeds, some berries, insects. Song poorly developed, weak chirps, chatterings, buzzes.

Family Ploceidae (weaverfinches). Small, stoutly built seedeaters, mostly 10 to 25 cm (4 to 10 in.), a few long-tailed species to 50 cm (males only). Short, stout, conical, pointed bills; gonys less than half length of upper bill; rictal bristles lacking; tarsus relatively short. Wings short to long, rounded to pointed, 10th primary present, much reduced; tail short to very long. Colours, patterns widely varied; sexes alike or unalike. About 156 species in all terrestrial habitats, mainly in Africa, also in Eurasia, Madagascar, Malaysia; a few (*e.g.,* the house sparrow, *Passer domesticus*) widely introduced elsewhere. Usually gregarious; food mainly seeds, some other vegetable matter, insects. Harsh, monotonous voices, song poorly developed.

Critical appraisal. The classification of the Passeriformes is currently in a state of flux. Of the several major works on the order since 1950, no two have agreed on the sequence of families, and each author has proposed innovations of his own. Ornithologists who are principally concerned with arranging museum displays, preparing national lists, or studying avian biology in the field may regard any change in classification as an unnecessary and unmitigated nuisance, and a few such workers have tried to prevent further change by having avian taxonomy formally "standardized." An attempt to do so at the XIVth International Ornithological Congress in 1966 failed in part because taxonomists opposed it vigorously and in part because its proponents could not agree on which of the several passerine sequences currently in use should be adopted.

Difficulties in preparing a lineal sequence

A main difficulty with any lineal sequence is the impossibility of making it reflect evolution's three-dimensional development, and it will be many years before enough

Comparison of Systems of Passeriform Classification

Encyclopædia Britannica	Wetmore, 1960	Peters Checklist	Storer, 1971
Order Passeriformes	**Passeriformes**	**Passeriformes**	**Passeriformes**
Suborder Eurylaimi	*Eurylaimi*	*Eurylaimi*	*Eurylaimi*
Family Eurylaimidae (broadbills)	Eurylaimidae	Eurylaimidae	Eurylaimidae
Suborder Tyranni	*Tyranni*	*Tyranni*	*Furnarii*
Superfamily Furnarioidea	Furnarioidea	Furnarioidea	
Family Dendrocolaptidae (woodcreepers)	Dendrocolaptidae	Dendrocolaptidae	Dendrocolaptidae (incl. Furnariidae)
Furnariidae (ovenbirds)	Furnariidae	Furnariidae	
Formicariidae (antbirds) (incl. Conopophagidae, part)	Formicariidae	Formicariidae	Formicariidae (incl. Conopophagidae, part)
	Conopophagidae (antpipits)	Conopophagidae	
Rhinocryptidae (tapaculos)	Rhinocryptidae	Rhinocryptidae	Rhinocryptidae
Superfamily Tyrannoidea	Tyrannoidea	Tyrannoidea	*Suborder Tyranni*
Family Cotingidae (cotingas)	Cotingidae	Cotingidae	Cotingidae
Pipridae (manakins)	Pipridae	Pipridae	Pipridae
Tyrannidae (tyrant flycatchers)	Tyrannidae	Tyrannidae	Tyrannidae
Oxyruncidae (sharpbill)	Oxyruncidae	Oxyruncidae	Oxyruncidae
Phytotomidae (plantcutters)	Phytotomidae	Phytotomidae	Phytotomidae
Pittidae (pittas)	Pittidae	Pittidae	
Xenicidae (New Zealand wrens)	Acanthisittidae (=Xenicidae)	Xenicidae	*Suborder Menurae*
Philepittidae (asities)	Philepittidae	Philepittidae	Atrichornithidae
			Menuridae
Suborder Menurae	*Menurae*	*Menurae*	*Suborder?*
Family Menuridae (lyrebirds)	Menuridae	Menuridae	Xenicidae
Atrichornithidae (scrub-birds)	Atrichornithidae	Atrichornithidae	Pittidae
			Philepittidae
Suborder Passeres	*Passeres*	*Passeres*	*Passeres*
Family Alaudidae (larks)	Alaudidae	Alaudidae	Palaeospizidae
Palaeospizidae (fossil only)	Palaeospizidae		Alaudidae
Hirundinidae (swallows)	Hirundinidae	Hirundinidae	Hirundinidae
Dicruridae (drongos)	Dicruridae	Motacillidae	Campephagidae
Oriolidae (Old World orioles)	Oriolidae	Campephagidae	Pycnonotidae
Corvidae (crows and jays)	Corvidae	Pycnonotidae	Irenidae
Callaeidae (wattlebirds)	Cracticidae	Irenidae	Laniidae
Grallinidae (mudnest builders)	Grallinidae	Laniidae (incl. Prionopidae)	Vangidae
Cracticidae (bellmagpies)	Ptilonorhynchidae	Vangidae (incl. Hyposittidae)	Bombycillidae
Ptilonorhynchidae (bowerbirds)	Paradisaeidae	Bombycillidae (incl. Ptilogonatidae)	Dulidae
Paradisaeidae (birds of paradise)	Paridae	Dulidae	Motacillidae
Paridae (titmice)	Sittidae	Cinclidae	Cinclidae
Certhiidae (creepers)	Hyposittidae (coral-billed nuthatch)	Troglodytidae	Troglodytidae
Sittidae (nuthatches)	Certhiidae	Mimidae	Mimidae
Climacteridae (Australian treecreepers)	Paradoxornithidae (=Panuridae)	Prunellidae	Prunellidae
Panuridae (bearded tits, parrotbills)	Chamaeidae	Muscicapidae	Muscicapidae
Chamaeidae (wrentit)	Timaliidae	Subfamily Turdinae	Subfamilies not listed, but family constituted basically as in Peters Checklist arrangement.
Timaliidae (babblers)	Campephagidae	Orthonychinae (log runners)	
Campephagidae (cuckoo-shrikes)	Pycnonotidae	Timaliinae	
Pycnonotidae (bulbuls)	Palaeoscinidae	Panurinae	
Palaeoscinidae (fossil only)	Chloropseidae (leafbirds, ioras)	Picathartinae (rockfowl)	
Irenidae (leafbirds, ioras, fairy bluebirds)	Cinclidae	Polioptilinae	
Cinclidae (dippers)	Troglodytidae	Sylviinae	
Troglodytidae (wrens)	Mimidae	Malurinae	
Mimidae (mockingbirds and allies)	Turdidae	Muscicapinae	
Turdidae (thrushes)	Zeledoniidae	Platysteirinae (wattle-eyes)	
	Sylviidae	Monarchinae (monarch flycatchers)	
Sylviidae (Old World warblers, incl. Regulidae)	Regulidae (kinglets)	Pachycephalinae	
Polioptilidae (gnatcatchers)	Muscicapidae	Aegithalidae (long-tailed tits)	
Pachycephalidae (whistlers)	Prunellidae	Remizidae (penduline titmice)	Aegithalidae
Maluridae (wren-warblers)	Motacillidae	Paridae	Climacteridae
Muscicapidae (Old World flycatchers)	Bombycillidae	Sittidae	Rhabdornithidae
Prunellidae (accentors)	Ptilogonatidae	Certhiidae	Certhiidae
Motacillidae (wagtails)	Dulidae	Rhabdornithidae (Philippine creepers)	Sittidae
Bombycillidae (waxwings)	Artamidae	Climacteridae	Paridae
Ptilogonatidae (silky flycatchers)	Vangidae	Dicaeidae	Remizidae
Dulidae (palm chat)	Laniidae	Nectariniidae	Dicaeidae
Hypocoliidae (hypocolius)	Prionopidae	Zosteropidae	Nectariniidae
Artamidae (wood-swallows)	Cyclarhidae	Meliphagidae	Zosteropidae
Vangidae (vanga shrikes)	Vireolaniidae	Emberizidae	Meliphagidae
Laniidae (shrikes)	Callaeidae	Subfamily Emberizinae (buntings)	Oriolidae
Prionopidae (helmet shrikes)	Sturnidae	Catamblyrhynchinae	Dicruridae
Sturnidae (starlings)		Cardinalinae (cardinal-grosbeaks)	
	Meliphagidae	Thraupinae	Callaeidae
Meliphagidae (honeyeaters)	Nectariniidae	Tersininae	Grallinidae
Nectariniidae (sunbirds)	Dicaeidae	Parulidae	Artamidae
Dicaeidae (flowerpeckers)	Zosteropidae	Drepanididae	Cracticidae
Zosteropidae (white-eyes)	Vireonidae	Vireonidae	Ptilonorhynchidae
Cyclarhidae (pepper-shrikes)	Coerebidae (honeycreepers)	Icteridae	Paradisaeidae
Vireolaniidae (shrike-vireos)	Drepanididae	Fringillidae	Corvidae
Vireonidae (vireos)	Parulidae	Estrildidae	Sturnidae
Drepanididae (Hawaiian honeycreepers)	Ploceidae	Ploceidae	Ploceidae
Parulidae (wood warblers)	Icteridae	Sturnidae	Estrildidae
Zeledoniidae (wrenthrush)	Tersinidae	Oriolidae	Fringillidae
Icteridae (New World orioles and allies)	Thraupidae	Dicruridae	Vireonidae
Tersinidae (swallow-tanager)	Catamblyrhynchidae	Callaeidae	Drepanididae
Thraupidae (tanagers)	Fringillidae	Grallinidae	Parulidae
Catamblyrhynchidae (plush-capped finch)		Artamidae	Emberizidae
Fringillidae (New World seedeaters)		Cracticidae	Icteridae
Carduelidae (goldfinches and allies)		Ptilonorhynchidae	
Estrildidae (waxbills)		Paradisaeidae	
Ploceidae (weaverfinches)		Corvidae	

evidence is in to satisfy all taxonomists. Most taxonomists place adjacent in the linear order those groups that they believe to be most closely related, but if three groups are believed to be equally interrelated, no linear sequence can express the relationship. The simplest solution for those who want a rigid, nonvarying sequence is to list all families alphabetically, and under each the genera and species alphabetically in turn. Even this system, which of course gives no indication of relationships, is not foolproof, for all scientific names are subject to change through the laws of priority as well as by inevitable changes in various taxa as more is learned about them. Hence it seems far better to retain the systems currently in use, leaving them fluid enough to be changed when changes are shown justifiable.

Since the late 19th century, when the many advances in taxonomic thought of the previous century began to crystallize and bear fruit, three main passerine sequences have dominated the world bird lists. The first, proposed originally by R.B. Sharpe of the British Museum in 1877, but based in part on the published and unpublished work of others, places the crows at the summit of avian evolution (hence at the end of a modern lineal sequence), ostensibly (although never so stated directly) on the basis of their alleged high intelligence. This sequence was adopted by the German ornithologist Ernst Hartert in his monumental *Die Vögel der paläarktischen Fauna,* published in 1903, and subsequently by most other European ornithologists.

The second sequence, placing the thrushes at the end, was in general usage, particularly in North America, until the late 1920s.

The third sequence, originally proposed in 1926 by two Americans, Alexander Wetmore and Waldron De-Witt Miller, but also based partly on the earlier work of others, places the crows, actually a rather generalized group, near the base of the oscine family tree, and in top place puts the so-called nine-primaried oscines, dominated by the seed-eating fringillids, because, as Wetmore explained in a 1960 paper (*A Classification for the Birds of the World*),

> this group is the modern expression of a main core or stem that through the earlier Tertiary periods has given rise to more specialized assemblages . . . . Further specialization is apparent in some parts of the existing fringilline assemblage that, if undisturbed, may lead to further differentiation.

This sequence immediately became standard for North American works and has remained so for these and certain international lists ever since, with only minor departures and rearrangements.

Within the suboscines, research is beginning to suggest that the three Old World families long included within the suborder Tyranni should be removed, leaving that suborder purely New World in distribution. It has been suggested that the asities (Philepittidae) may be more closely related to the broadbills, with which they share some features of the syrinx and sternum, than to the New World tyrannoid groups. The relationships of the asities, pittas, and New Zealand wrens will remain uncertain until further study can provide a better understanding of their biology and anatomy.

Nearly all recent classifications have grouped the eight to 10 families of New World suboscines in a suborder, Tyranni, based on syringeal features studied by the 19th-century anatomist A.H. Garrod. An American ornithologist, P.L. Ames, following a re-examination of the syringeal anatomy, suggested in 1965 that one superfamily of the suborder, Furnarioidea (the antbird-ovenbird group), should not be united to the other, Tyrannoidea (tyrant flycatchers and allies), on the basis of the syrinx but should be elevated to subordinal rank, a view supported by features of the pterylosis and sternal and cranial osteology. This suggestion, made initially in 1907 by a British anatomist, W.P. Pycraft, but without supportive evidence, was taken up in a classification by R.W. Storer in 1971 (see the Table).

As more suboscines are studied in detail, some genera may be found to have been misclassified; particularly vulnerable are certain forms in the woodcreeper–ovenbird (Dendrocolaptidae–Furnariidae), antbird–tapaculo (Formicariidae–Rhinocryptidae), and cotinga–manakin–flycatcher (Cotingidae–Pipridae–Tyrannidae) complexes. A group of small South American ground birds known as gnateaters and antpipits was separated from the antbirds in 1882 as a family, Conopophagidae, and generally recognized for nearly 90 years. Careful study in 1968, however, revealed that the family was an artificial one; one genus (*Conopophaga*) has been returned to the Formicariidae, the other (*Corythopis*) placed in the Tyrannidae. Recent studies of egg-white proteins by American biologist C.G. Sibley and colleagues suggest that the palm chat (*Dulus*) may be closer to the weaverfinches than to the waxwings; that the Cape sugarbird (*Promerops*) may be a starling rather than a honeyeater; and that the phainopepla (*Phainopepla*) may be most closely related to the solitaires (*Myadestes*), which are currently placed in the thrush family. A number of widely accepted oscine families invite similar examination. Only additional study will solve these problems and (it is hoped) provide a better understanding of the position of aberrant genera, such as *Rupicola* and *Menura.*

Within the oscines, one of the greatest problems is a satisfactory delineation and arrangement of the many superficially similar groups. Most taxonomists agree that the oscines contain three large groups (see the Table): (1) the crows, Old World orioles, birds of paradise, and allies; (2) the thrushes, babblers, Old World flycatchers, Old World warblers, kinglets, and allies; and (3) the finches, icterids, tanagers, and allies. Authorities disagree on how to subdivide these groups into "manageable" sections, precisely where to draw the lines between some families, and even whether to assign them familial or lesser taxonomic rank. Other groups, currently assigned family rank, are still of moot affinities and defy all efforts to place them in a linear sequence completely acceptable to everyone; among these are the larks, swallows, titmice, wrens, dippers, mimid thrushes, honey eaters, waxwings, and the palm chat.

As present-day students are delving more deeply into and reassessing the anatomical evidence on which most passerine families were erected a century ago, they are finding that much oscine classification was based originally on a few basic characters, weighted too strongly by the opinions of individual taxonomists. Bringing to bear such existing new lines of evidence as ethology, serology, parasitology, and protein analysis is helping to clear some of the long unchallenged beliefs that have clouded the main issues. It is hoped that these developments will encourage taxonomists to abandon some of their tenacious opinions—*e.g.,* that the crows represent the apex of passerine evolution. They should then be able to draw familial lines at the point at which the evidence indicates they should be, regardless of the relative size of each group.

(M.H.C./O.L.A.)

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Podicipediformes. R.S. PALMER (ed.), *Handbook of North American Birds,* vol. 1 (1962), provides an up-to-date summary of information on North American species; K.E.L. SIMMONS, "Studies on Great Crested Grebes," *Avicult. Mag.,* 61:3–13, 93–102, 131–146, 181–201, 235–253, 294–316 (1955), is a classic study of grebe behaviour; R.W. STORER, "The Behavior of the Horned Grebe in Spring," *Condor,* 71:180–205 (1969), contains comparative material on several species, and "The Patterns of Downy Grebes," *Condor,* 69:469–478 (1967), is concerned with the colour patterns of grebe chicks. The same author's "Courtship and Mating Behavior and the Phylogeny of the Grebes," *Proc. 13th Int. Orn. Congr.,* pp. 562–569 (1963), and "Evolution in the Diving Birds," *Proc. 12th Int. Orn. Congr.,* pp. 694–707 (1960), provide insight into the evolution and classification of grebes.

Procellariiformes. Among the general works of value W.B. ALEXANDER, *Birds of the Ocean,* 3rd ed. (1955), remains a standard identification handbook; A.C. BENT, *Life Histories of North American Petrels and Pelicans and Their Allies* (1922); R.A. FALLA *et al., Birds of New Zealand* (1966), deals with some 56 tubinare species; J. FISHER and R.M. LOCKLEY, *Sea-birds* (1954), an introduction to the natural history of the sea-birds of the North Atlantic; R.M. LOCKLEY, *The Island* (1969), includes life-histories of Manx shearwaters and storm petrels, and homing experiments; *Man Against Nature* (1970), descriptions of New Zealand tubinares and muttonbirding; R.C. MURPHY, *Oceanic Birds of South America,* 2 vol. (1936), also many original papers by this noted field worker; R.S. PALMER (ed.), *Handbook of North American Birds,* vol. 1 (1962), contains much information about the species recorded from North America; V. SERVENTY, *A Continent in Danger* (1966), a description of muttonbirding and banding of shearwaters; H.F. WITHERBY (ed.), *The Handbook of British Birds,* vol. 4 (1949), detailed characters and life histories of tubinares recorded in the British Isles.

Notable monographs include: W.B. ALEXANDER *et al.,* "The Families and Genera of the Petrels and Their Names," *Ibis,* 107:401–405 (1965), a taxonomical reorganization of the order Procellariiformes; P.E. DAVIS, "The Breeding of the Storm Petrel," *Br. Birds,* 50:85–101, 371–84 (1957), follow-up of an earlier study by Lockley; J. FISHER, *The Fulmar* (1952), a classic monograph; W.S. JAMESON, *The Wandering Albatross* (1958), a readable account of the biology of the largest albatross; R.M. LOCKLEY, *Shearwaters* (1961), an edition of a classic monograph first published in 1942; R.C. MURPHY and L.S. MOWBRAY, "New Light on the Cahow, *Pterodroma cahow,*" *Auk,* 68:226–80 (1951), rediscovery of the cahow at Bermuda; B. NELSON, *Galapagos: Islands of Birds* (1968), contains an extensive section on the waved albatross; M.K. ROWAN, "The Greater Shearwater *Puffinus gravis* at Its Breeding Grounds," *Ibis,* 94:97–121 (1952); J. WARHAM, "The Biology of the Giant Petrel *Macronectes giganteus,*" *Auk,* 79:139–60 (1962); D.B. WINGATE, "Discovery of Breeding Black-Capped Petrels on Hispaniola," *Auk,* 81:147–59 (1964).

Sphenisciformes. JOHN SPARKS and TONY SOPER, *Penguins* (1967); and BERNARD STONEHOUSE, *Penguins* (1968), are well illustrated popular books. R.C. MURPHY, *Oceanic Birds of South America,* 2 vol. (1936), is a classic work with extensive information on 10 species of penguins. J.T. EMLEN and R.L. PENNEY, "Distance Navigation in the Adélie Penguin," *Ibis,* 106:417–431 (1964), is a short technical paper concerned with long-range movements. R.L. PENNEY, "Territorial and Social Behaviour in the Adélie Penguin," the "Antarctic Research Series," vol. 12, *Antarctic Bird Studies,* pp. 83–131 (1968), reviews many studies of the behaviour of this species. JEAN PREVOST, *Écologie du Manchot empereur, Aptenodytes forsteri Gray* (1961), is a detailed study, in French, of the largest penguin species. Also of interest is L.E. RICHDALE, *Sexual Behavior in Penguins* (1951). G.G. SIMPSON, "Fossil Penguins," *Bull. Am. Mus. Nat. Hist.,* 87:1–99 (1946), is a basic work on the evolution of penguins.

Pelecaniformes. R.S. PALMER (ed.), *Handbook of North American Birds,* vol. 1 (1962), detailed accounts of the biology of all species of North American pelecaniforms; R.C. MURPHY, *Oceanic Birds of South America,* 2 vol. (1936), a classic work that includes excellent accounts of the guano birds of the Peru Current; D. LACK, *Ecological Adaptations for Breeding in Birds* (1968), an analysis that includes a useful discussion of pelecaniforms; G.F. VAN TETS, "A Comparative Study of Some Social Communication Patterns in the Pelecaniformes," *Amr. Orn. Union, Orn. Monogr. No. 2* (1965), a general discussion of behaviour and relationships of the different groups; B. NELSON, *Galapagos: Islands of Birds* (1968), readable and authoritative accounts of the ecology and behaviour of tropical pelecaniforms, especially boobies and frigatebirds; P. BRODKORB, "Catalogue of Fossil Birds: Part 1 (Archaeopterygiformes through Ardeiformes)," *Fla. St. Mus. Bull.* 7, pp. 179–293 (1963), a list of the fossil pelecaniform birds.

Anseriformes. G.L. ATKINSON-WILLES (ed.), *Wildfowl in Great Britain* (1963), a regional study of distribution and populations; K.M. BAUER and U.N. GLUTZ VON BLOTZHEIM, *Handbuch der Vögel Mitteleuropas,* vol. 2 and 3, *Anseriformes* (1968–69), much wider coverage than the name implies, the most complete source book on the biology of Holarctic waterfowl; H.J. BOYD, G.V.T. MATTHEWS, and M.A. OGILVIE (eds.), *Wildfowl,* formerly *Annual Reports of the Wildfowl Trust* (1950–73), 24 volumes, containing articles on all aspects of the biology of wildfowl; J. DELACOUR, *The Waterfowl of the World,* 4 vol. (1954–64), monographic treatment of all species with coloured plates by Peter Scott; H.J. FRITH, *Waterfowl in Australia* (1967), the only book devoted to waterfowl of the Southern Hemisphere; H.A. HOCHBAUM, *The Canvasback on a Prairie Marsh,* 2nd ed. (1959), a classic species monograph, *Travels and Traditions of Waterfowl* (1955), a review of waterfowl migration and orientation; P.A. JOHNSGARD, *Handbook of Waterfowl Behaviour* (1965), an inventory of behaviour patterns, especially social and sexual, *Waterfowl: Their Biology and Natural History* (1968), contains photographs, many coloured, of nearly every species; F.H. KORTRIGHT, *The Ducks, Geese and Swans of North America* (1942), a thorough treatment, with excellent illustrations; J.P. LINDUSKA (ed.), *Waterfowl Tomorrow* (1964), authoritative articles on all aspects of waterfowl conservation, management, and research; J.G. MILLAIS, *The Natural History of the British Surface-feeding Ducks* (1902), an early classic, still of value, *British Diving Ducks* (1913), two-volume work with much useful information; J.C. PHILLIPS, *The Natural History of the Ducks,* 4 vol. (1922–26), a classic monograph covering many aspects of waterfowl biology; P. SCOTT, *A Coloured Key to the Wildfowl of the World,* rev. ed. (1972), contains coloured plates of all species of the Anatidae, with details of distribution and aids to identification.

Ciconiiformes. R.P. ALLEN, *The Flamingos: Their Life History and Survival* (1956), a survey of the family; W.J. BOCK, "A Generic Review of the Family Ardeidae (Aves)," *Am. Mus. Novit.,* no. 1779 (1956), a discussion of the taxonomy of the herons; L.H. BROWN, *Mystery of the Flamingos* (1959), a description of the natural history of the East African species; F.A. LOWE, *The Heron* (1954), a discussion of the natural history of the European grey heron (*Ardea cinerea*); A.J. MEYERRIECKS, *Comparative Breeding Behaviour of Four Species of North American Herons* (1960), a valuable contribution to knowledge of behaviour in the group; R.S. PALMER (ed.), *Handbook of North American Birds,* vol. 1 (1962), detailed accounts of 22 species of the order; A.L. THOMSON (ed.), *A New Dictionary of Birds* (1964), short articles on individual ciconiiforms.

Falconiformes. R. AUSTING, *The World of the Red-Tailed Hawk* (1964), a photographic essay on the daily life of this common North American buteo; A.C. BENT, "Life Histories of North American Birds of Prey," *Bull. U.S. Natn. Mus.*, vol. 167, pt. 1 (1937), and vol. 170, pt. 2 (1938), a review by species of the natural history of all of the falconiforms and strigiforms (owls) found in North America; L.H. BROWN, *Eagles* (1955; shortened version, 1970), an account of the biology of these impressive birds, written for the layman, and with D. AMADON, *Hawks, Eagles and Falcons of the World* (1969), extensive information on the natural history of every species of the falconiforms; L.H. BROWN, *African Birds of Prey* (1971), an ecological study of falconiforms and strigiforms on the African continent; P.E. BROWN and G. WATERSTON, *The Return of the Osprey* (1962), a detailed account of the natural history of the osprey; J.J. and F.C. CRAIGHEAD, *Hawks, Owls and Wildlife* (1956), a detailed study of the food ecology of raptorial birds and their effect on prey populations; M.L. GROSSMAN and J. HAMLET, *Birds of Prey of the World* (1964), a heavily illustrated work dealing with both falconiforms and owls; J.J. HICKEY (ed.), *Peregrine Falcon Populations* (1969), a report of a symposium to determine the cause of the decline of the falconiform populations, especially that of the peregrine; H.K. SWANN, *Monograph of the Birds of Prey (Order Accipitres),* vol. 1 (1930), vol. 2 (1945), a classic work on the biology of the falconiforms of the world, with many fine colour plates.

Galliformes (works arranged by family). *Megapodiidae:* H.J. FRITH, *The Mallee-Fowl: The Bird That Builds An Incubator* (1962), is an extensive study of the breeding habits of *Leipoa ocellata.*

Cracidae: A.F. SKUTCH, "Habits of the Chestnut-winged Chachalaca," *Wilson Bull.*, 75:262–269 (1963), presents information on the breeding habits of *Ortalis garrula.* C. VAURIE, "Taxonomy of the Cracidae (Aves)," *Bull. Am. Mus. Nat. Hist.*, 138:133–259 (1968); and F. VUILLEUMIER, "Relationships and Evolution Within the Cracidae (Aves, Galliformes)," *Bull. Mus. Comp. Zool. Harv.*, 134:1–27 (1965), are two analyses of the taxonomic relationships within this family.

Tetraonidae: A.W. BOBACK, *Das Auerhuhn* (1952), is a short monograph on the capercaille; and A.W. BOBACK and D.W. MULLER-SCHWARZE, *Das Birkhuhn* (1968), is a study of the black grouse. G.L. GIRARD, "Life History, Habits, and Food of the Sage Grouse, *Centrocercus urophasianus* Bonaparte," *Univ. Wyo. Publs.*, 3:1–56 (1937), provides insight into the natural history of this grassland species. G. BUMP *et al., The Ruffed Grouse: Life History, Propagation, Management* (1947), is a classic study of this woodland grouse.

Phasianidae: C.W. BEEBE, *A Monograph of the Pheasants,* 4 vol. (1918–22), is a classic work on the pheasants, magnificently illustrated with colour plates. A revised, less technical edition was published in two volumes under the title *Pheasants: Their Lives and Homes* (1926) and later in one volume (1936). J.T. DELACOUR, *The Pheasants of the World* (1951), is a fully illustrated treatment of the group, leaning heavily toward aviculture. H.L. STODDARD, *The Bobwhite Quail: Its Habits, Preservation and Increase* (1931), is an important book on this game species. P. WAYRE, *A Guide to the Pheasants of the World* (1969), is an extensively illustrated work with emphasis on conservation and on breeding in captivity.

Numididae: A.R. LEE, "The Guineafowl," *Fmrs' Bull. U.S. Dep. Agric. 1391* (1924), presents information on the guinea fowl, from an avicultural standpoint.

Meleagrididae: H.S. MOSBY and C.O. HANDLEY, *The Wild Turkey in Virginia* (1943), is a broad study of the natural history of *Meleagris gallopavo.* A.W. SCHORGER, *The Wild Turkey: Its History and Domestication* (1966), is primarily a historical treatise.

Opisthocomidae: C.W. BEEBE, "A Contribution to the Ecology of the Adult Hoatzin," *Zoologica,* 1:45–66 (1909); and *Tropical Wildlife in British Guiana,* pp. 155–182 (1917), present information from the author's field observations of the hoatzin.

Gruiformes. A.C. BENT, "Life Histories of North American Marsh Birds," *Bull. U.S. Natn. Mus.* no. 135 (1926), contains long and comprehensive accounts of North American cranes and rails and the limpkin; H.F. WITHERBY *et al.* (eds.), *The Handbook of British Birds,* vol. 4 and 5 (1940–41), provides excellent accounts of European cranes, rails, and bustards, with coloured plates; R.P. ALLEN, *The Whooping Crane* (1952), is the classic monograph on this endangered species, with a brief descriptive and distributional treatment of the cranes of the world; L.H. WALKINSHAW, "The Sandhill Cranes," *Bull. Cranbrook Inst. Sci.* no. 29 (1949), contains detailed information on this species; H.E. HOWARD, *A Waterhen's Worlds* (1940), is a monograph of the common gallinule or moorhen, with emphasis on behaviour, one of the early classics in ethology; S. KEITH, C.W. BENSON, and M.P. STUART IRWIN, "The Genus *Sarothrura* (Aves, Rallidae)," *Bull. Am. Mus. Nat. Hist.*, vol. 143, art. 1 (1970), is a study of this genus of African rails, with colour plates of the birds, diagrams of vocalizations, and a general treatment of rail voices; H.T. HENDRICKSON, "A Comparative Study of the Egg White Proteins of Some Species of the Avian Order Gruiformes," *Ibis,* 3:80–91 (1969), contains an exhaustive listing of the taxonomic characters for the order.

Charadriiformes. General works include C.G. LOW, *The Literature of the Charadriiformes from 1894–1928,* 2nd rev. and enl. ed. (1931); and H. SEEBOHM, *The Geographical Distribution of the Family Charadriidae* (1888), a classical, well-illustrated monograph that includes the Scolopacidae. Form and function are treated in J. BEDARD, "Adaptive Radiation in Alcidae," *Ibis,* 111:189–198 (1969), on feeding adaptations; F.A. HARTMAN, "Locomotor Mechanisms of Birds," *Smithson. Misc. Collns.*, vol. 143, no. 1 (1961); R. MEINERTZHAGEN, "The Speed and Altitude of Bird Flight," *Ibis,* 97:81–117 (1955); and R.W. STORER, "Evolution in the Diving Birds," *Proc. XII Int. Orn. Congr.*, pp. 694–707 (1960), on locomotor adaptations; R.L. ZUSI, "Structural Adaptations of the Head and Neck in the Black Skimmer (*Rynchops nigra* Linnaeus)," *Publs. Nuttall Orn. Club,* no. 3 (1962), a technical discussion of feeding behaviour and adaptations.

Many short accounts of the natural history occur in regional books. A.C. BENT has written on the life histories of the North American diving birds, gulls and terns, and shorebirds, published in the *Bull. U.S. Natn. Mus.*, no. 107, 113, 142, and 146 (1919, 1921, 1927, and 1929). R.H. DRENT, "Breeding Biology of the Pigeon Guillemot, *Cepphus columba,*" *Ardea,* 53:99–160 (1965); R.C. MURPHY, *Oceanic Birds of South America,* vol. 2 (1936), a readable discussion of many southern species; D. NETHERSOLE-THOMPSON, *The Greenshank* (1951), a very complete life history account; R.S. PALMER, "A Behavior Study of the Common Tern (*Sterna hirundo hirundo* L.)," *Proc. Boston Soc. Nat. Hist.*, 42:1–119 (1941); N. TINBERGEN, *The Herring Gull's World* (1953), an exceptionally clearly written behavioral study; G.D. STOUT (ed.), *The Shorebirds of North America* (1967), excellent illustrations and detailed species accounts.

The major sources of general information on charadriiform fossils are P. BRODKORB, "Catalogue of Fossil Birds: Part 3 (Ralliformes, Ichthyornithiformes, Charadriiformes)," *Bull. Fla. St. Mus. Biol. Sci.*, 2:99–220 (1967), and K. LAMBRECHT, *Handbuch der Palaeornithologie,* 8 vol. (1933).

Taxonomic works include W.J. BOCK, "A Generic Review of the Plovers (Charadriinae, Aves)," *Bull. Mus. Comp. Zool. Harv.* vol. 118, no. 2 (1958), on the function and value of traditional taxonomic characters; J.R. JEHL, JR., "Relationships in the Charadrii (Shorebirds): A Taxonomic Study Based on Color Patterns of the Downy Young," *Mem. S. Diego Soc. Nat. Hist. 3* (1968), a discussion of a taxonomic character not previously stressed; P.R. LOWE, "An Anatomical Review of the 'Waders' (Telmatomorphae) with Special Reference to the Families, Subfamilies and Genera Within the Suborders Limicolae, Grui-Limicolae, and Lari-Limicolae," *Ibis,* 13th ser., 1:712–771 (1931), an important classification based mainly on anatomy; M. MOYNIHAN, "A Revision of the Family Laridae (Aves)," *Am. Mus. Novit.*, no. 1928 (1959), the basis for many classifications of gulls and terns; J.L. PETERS, *Check-list of Birds of the World,* vol. 2 (1934), the distribution of all species; R. RIDGWAY, "The Birds of North and Middle America," *Bull. U.S. Natn. Mus.*, no. 50, pt. 8 (1919), taxonomic characters of the order, families, genera, and species.

Columbiformes. For accounts of the natural history of columbiform birds, the reader should consult D. GOODWIN, *Pigeons and Doves of the World* (1967); for the biology of an abundant European species, R.K. MURTON, *The Wood Pigeon* (1965); for Australian species, H.J. FRITH, "Notes on the Pigeons of the Richmond River, N.S.W.," *Emu,* 52:89–99 (1952); for North American pigeons, A.C. BENT, "Life Histories of North American Gallinaceous Birds," *Bull. U.S. Natn. Mus. 162,* pp. 353–477 (1932); and for neotropical species, A.F. SKUTCH, "Life Histories of Central American Pigeons," *Wilson Bull.*, 76:211–247 (1964). W.M. LEVI, *The Pigeon,* rev. ed. (1957), provides a complete discussion of the care and breeding of domestic pigeons. A.W. SCHORGER, *The Passenger Pigeon* (1955), is an exceptionally thorough account of the natural history, economics, and demise of this once abundant North American pigeon. In addition to Goodwin's book, the taxonomy of pigeons is discussed in R.F. JOHNSTON, "Taxonomy of Pigeons," *Condor,* 64:69–74 (1962); D. GOODWIN "Taxonomy of the Genus *Columba,*" *Bull. Br. Mus. Nat. Hist. (Zool.)*, no. 6, pp. 1–23 (1959); and E.W. GIFFORD, "Taxonomy and Habits of Pigeons." *Auk,* 58:239–245 (1941).

Psittaciformes. Popular works on parrots and related birds include J.M. FORSHAW, *Australian Parrots* (1969); and W.R. EASTMAN, JR. and A.C. HUNT, *Parrots of Australia* (1966), both large, lavish books describing and depicting in colour all Australian parrots. H. BATES and R. BUSENBARK, *Parrots and*

Related Birds (1959); and the DUKE OF BEDFORD, *Parrots and Parrot-like Birds* (1954), treat all of the parrots, with particular reference to aviculture. H.S. ZIM, *Parakeets* (1953), is devoted to the aviculture of the budgerigar. W.C. DILGER, "The Behavior of Lovebirds," *Scient. Am.*, 206:88–98 (1962), is an excellent summary of the author's research into the behaviour and evolution of *Agapornis*.

Among the more technical works: H. VON BOETTICHER, *Papageien* (1959), is a 116-page booklet, in German, containing 10 pages of introductory information, followed by a new classification based on external morphology, especially of the cere, with all species described briefly. A. NEWTON, *A Dictionary of Birds* (1893–96), contains a seven-page summary of parrot biology, as known at the end of the last century. J.L. PETERS, *Check-List of Birds of the World*, vol. 3 (1937), presents a classification of all parrots down to subspecies, with information on geographic ranges and scientific names. A. REICHENOW, *Vogelbilder aus fernen Zonen-Papageien*, 2nd ed. (1955), includes colour paintings and brief descriptions, in German, of almost all parrots.

Cuculiformes. E. BALDAMUS, *Das Leben der europäischen Kuckucke* (1892); and W. VON CAPEK, "Beiträge zur Fortpflanzungsgeschichte des Kuckucks," *Orn. Jb.*, 7:42–72, 102–117, 146–157, 165–183 (1896), are two classic works, in German, about the common cuckoo. E.P. CHANCE, *The Cuckoo's Secret* (1922), is an important and highly readable book about the common cuckoo. D.E. DAVIS, "Social Nesting Habits of the Smooth-Billed Ani," *Auk*, 57:179–218 (1940), presents the results of extensive field work on anis. H. FRIEDMANN, *The Parasitic Cuckoos of Africa* (1948); "Evolutionary Trends in the Avian Genus *Clamator*," *Smithson. Misc. Collns.*, 146:1–127 (1964), and "The Evolutionary History of the Avian Genus *Chrysococcyx*," *Bull. U.S. Natn. Mus. 265* (1968), are technical papers on Old World tropical cuckoos. A.C. BENT, "Life Histories of North American Cuckoos, Goatsuckers, Hummingbirds and Their Allies," *ibid. 176* (1940), provides readable accounts of the natural history of North American cuckoos and of some Asiatic *Cuculus* species. R.E. MOREAU, "A Contribution to the Biology of the Musophagiformes, the So-Called Plantain-Eaters," *Ibis*, 2: 639–671 (1938), is one of the few detailed studies of musophagids.

Strigiformes. O.L. AUSTIN, *Birds of the World*, ed. by H.S. ZIM (1961), presents an excellent general account of owls, with fine colour illustrations that include some rarely seen species. A.C. BENT, "Life Histories of North American Birds of Prey, Part II," *Bull. U.S. Natn. Mus. 170* (1938; reissued in 1961 as vol. 2 of series of same title), contains complete accounts of the natural history of North American owls. J.J and F.C. CRAIGHEAD, *Hawks, Owls and Wildlife* (1956, reissued in pa. 1969), is a classic in the ecology of raptorial birds. M.L. GROSSMAN and J. HAMLET, *Birds of Prey of the World* (1964), provides range maps for each species, good photographs, and a reading list by geographical region and by genus. G.P. DEMENTIEV *et al.*, *Birds of the Soviet Union*, vol. 1 (Eng. trans. 1966; orig. pub. in Russian, 1951), contains an extensive and well-illustrated section on Eurasian owls. E.C.S. BAKER, *Fauna of British India (Birds)*, vol. 4 (1927), includes an excellent discussion of the anatomical traits used in the classification of owls. Of a more specialized nature: R.S. PAYNE and W.H. DRURY, JR., "Marksman of the Darkness," *Nat. Hist.*, vol. 67, pp. 316–323 (June 1958), describes the discovery of the hearing ability of the barn owl. D.J. LIGON, "The Biology of the Elf Owl, *Micrathene Whitneyi*," *Misc. Publs. Mus. Zool. Univ. Mich. No. 136* (Feb. 1968), is a detailed study of the life history of the smallest owl, compared with other owls. A. ROSS, "Ecological Aspects of the Food Habits of Insectivorous Screech-owls," *Proc. West. Fdn. Vertebr. Zool.* vol. 1, no. 6, pp. 301–344 (1969), is a definitive correlation between body size and prey size, one of the few such papers in which stomach contents are correctly identified. L.W. WALKER, "Nocturnal Observations of Elf Owls," *Condor*, vol. 45, no. 5, pp. 165–167 (Sept.-Oct. 1943), with excellent photography and concise scientific writing. B.E. SMYTHIES, *The Birds of Burma*, 2nd ed. rev. (1953), and *The Birds of Borneo* (1960), provide the best and most reliable information on the habits and calls of Asian owls.

Caprimulgiformes. The natural history of North American nightjars is extensively discussed by A.C. BENT, "Life Histories of North American Cuckoos, Goatsuckers, Hummingbirds and Their Allies," *Bull. U.S. Natn. Mus. 176* (1940); that of European nightjars in D.A. BANNERMAN and G.E. LODGE, *The Birds of the British Isles*, vol. 4 (1955). F. HAVERSCHMIDT, *The Birds of Surinam* (1968), includes much information on the South American caprimulgiforms. J.P. CHAPIN, "Birds of the Belgian Congo, Part II," *Bull. Am. Mus. Nat. Hist.*, vol. 75 (1939), contains readable accounts of the natural history of several tropical nightjars. D.W. SNOW, "The Natural History of the Oilbird, *Steatornis caripensis*, in Trinidad, W.I.," *Zoologica*, vol. 46, pp. 27–48 (1961) and vol. 47, pp. 199–222 (1962), provides an excellent and readable account of the natural history of that species. The information on nightjar hibernation is scattered among a number of short journal articles, but the reader will find interesting summaries in E.C. JAEGER, "Further Observations on the Hibernation of the Poorwill," *Condor*, vol. 51, no. 3, pp. 105–109 (1949); and J.T. MARSHALL, JR., "Hibernation in Captive Goatsuckers," *Condor*, vol. 57, no. 3, pp. 129–134 (1955). A review of the later work in this field is presented by G.T. AUSTIN and W.G. BRADLEY, "Additional Responses of the Poorwill to Low Temperatures," *Auk*, vol. 86, no. 4, pp. 717–725 (1969).

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Birth Control

The term birth control refers to the volitional control of the number and spacing of children in a family. It encompasses the wide range of rational and irrational methods that have been used in the attempt to regulate human fertility, as well as the response of individuals and of groups within society to the choices offered by such methods. Birth control has been and remains controversial. The U.S. reformer Margaret Sanger coined the phrase in 1914–15 and, like the social movement she founded, the term has been caught up in a quest for acceptance, generating many synonyms: family planning, planned parenthood, responsible parenthood, voluntary parenthood, contraception, fertility regulation, and fertility control.

Human reproduction involves a range of activities and events, from sexual intercourse through birth, and depends as well on a series of physiological interactions, such as the timing of ovulation within the menstrual cycle. The visible events are central to the transmission of life and have been subject to social and religious control. The invisible factors in human reproduction gave rise early on to speculation and in modern times have become the topic of scientific investigation and manipulation. New knowledge relevant to birth control has diffused at different rates through various social groups and has not always been available to those with the greatest need. Hence, the conflicts and controversies surrounding birth control have been complex and impassioned. The disagreement over birth control arises in part from the debate over what is natural and what is artificial (and, to some, unacceptable). For information on human reproduction in general see REPRODUCTION AND REPRODUCTIVE SYSTEMS: *The human reproductive system* and *Human reproduction from conception to birth.*

This article is divided into the following sections:

NATURAL FERTILITY

At first glance the species *Homo sapiens* appears to have low potential for reproduction. Puberty begins late, pregnancy is long, normally only one baby is delivered at a time, and lactation can continue for several years. Yet on the global level the human race now experiences 1,000,000 more births than deaths every five days, and a large percentage of the world's population lives in urban areas, often at extremely high population densities. In experiments, when mammals are placed in crowded conditions the age of sexual maturity rises, the interval between pregnancies increases, and infant mortality jumps, leading to slower growth in the population. Among human beings in analogous crowded conditions, however, in the absence of artificial birth control the opposite situation arises.

In many cases ovulation does not take place in the first several cycles after the onset of menstruation (the menarche). Once a woman is fertile, social factors determine whether she is exposed to the opportunity to become pregnant. In preindustrial Britain, couples were expected to form their own nuclear group upon marriage, and many a first-time bride was in her later 20s. By contrast, in contemporary Third World societies that encourage extended families, girls often marry in the early teens.

Lactation's control over fertility

In all mammals whose reproduction is not tied to seasonal changes, physiological mechanisms ensure the optimum spacing of pregnancies. In *Homo sapiens,* as in other primates, breast-feeding provides the basis for nature's own method of birth control. In the few remaining societies of hunters and gatherers, whose way of life may represent the conditions under which most of human evolution took place, women nurse their babies frequently and ovulation and menstruation are suppressed for two to three years after birth. Nomadic women of the *!Kung,* a group of the San people of southern Africa, use no contraceptives but have a mean interval between births of 44 months and an average of four or five deliveries in a fertile lifetime. Modern methods of birth control substitute for the control over fertility once provided by lactation and permit a degree of control over human reproduction not previously available.

The combination of high infant mortality with relatively low fertility associated with traditional patterns of breast-feeding kept population growth in preagricultural human societies virtually static. Ten thousand years ago the world's population may have stood at 10,000,000. Since that time natural restraints on human reproduction have broken down at an accelerating pace. By the beginning of the Christian Era the world's population was perhaps 300,000,000. In the mid-1980s it passed the 5,000,000,000 mark. Since the Industrial Revolution, and with intensely increasing pressure in the past century, both individuals and societies have had to make important decisions about the use of birth control.

HISTORY OF BIRTH CONTROL

Methods. Written records of birth control methods survive from ancient times. Methods are mentioned among the various formulas and remedies recorded in the Ebers papyrus, a compilation of Egyptian medical texts dating from 1550 BC. Classical writers, including Pliny the Elder, Pedanius Dioscorides (*De materia medica, c.* AD 77), and Soranus of Ephesus (*On Midwifery and the Diseases of Women, c.* AD 100), refer to contraception and abortion. Several authors from the flowering of Arabic medicine in the 10th century mention contraception, notably ar-Rāzī (Rhazes; *Quintessence of Experience*), Ali ibn Abbas (*The Royal Book*), and Avicenna (Ibn Sīnā; *The Canon of Medicine*). The methods recommended by these early commentators fall into three groups: the reasonable but probably ineffective, such as wiping out the vagina after intercourse (Soranus); the reasonable and perhaps effective, such as using honey, alum, or lactic acid as spermicidal barriers (Ebers papyrus, Dioscorides, and Soranus); and the mystical and manifestly ineffective, for example suggesting that the woman jump backward seven times immediately after coitus (Soranus).

Ancient records of birth control methods

By 1900 all the methods of birth control now in use, with the exception of oral contraceptives, were understood and available in Europe and North America. The first to be developed was the condom. Folklore attributes the invention to a Dr. Condom, said to have been alarmed by the number of illegitimate offspring of Charles II of England. It is more likely, however, that the name derives from the Latin

Invention of the condom

condus, for receptacle. The earliest published description is by the Italian anatomist Gabriel Fallopius in 1564. The first condoms were made from animal intestines and for the most part were used to prevent sexual transmission of disease. When Charles Goodyear discovered the process for vulcanization of rubber in 1839 he initiated a revolution in contraception, as well as transport, and condoms have been a popular choice for birth control since the second half of the 19th century. The use of vaginal barriers (diaphragms and caps, which are commonly used with spermicides) was recorded by the German physician F.A. Wilde in 1823. The medical forerunner of the intrauterine device was the stem pessary, first described and illustrated in 1868. By 1909, Richard Richter, a practitioner from near Breslau (Wrocław in present-day Poland), had described most of the advantages and disadvantages of this method of birth control. Vasectomy, or male sterilization, was practiced in the 19th century, and the first female sterilization by surgical occlusion of the fallopian tubes was performed by a U.S. surgeon in Ohio in 1881. The surgical techniques of induced abortion in use today were also known in the 19th century. In the 1860s an Edinburgh gynecologist, James Young Simpson, described a procedure for "dry cupping" the uterus. The procedure adumbrated vacuum aspiration, a method commonly used today for performing legal abortion.

Ovulation and the menstrual cycle

The fact that conception was more likely to take place during certain phases of the menstrual cycle than others was suspected by classical authors. Adam Raciborski, a Paris physician, noted in 1843 that brides married soon after their menstruation often conceived in that cycle, while if the wedding occurred later in the cycle they commonly had another period before pregnancy occurred. Hermann Knaus in Austria (1929) and Kyūsaku Ogino in Japan (1930) independently and correctly concluded that ovulation occurs 14 days prior to the next menstruation. In 1964 an Australian medical team, John and Evelyn Billings, showed how women could monitor changes in their cervical mucus and learn to predict when ovulation would occur.

"The greatest invention some benefactor can give mankind," wrote Sigmund Freud in the early years of the 20th century, "is a form of contraception which does not induce neurosis." Many of the elements to meet the goal of a new, more acceptable form of contraception were present about the time of World War I, yet two generations were to reach maturity before those elements were exploited. The role of hormones from the ovary was understood early in the 20th century by Walter Heape and John Marshall. The first extract of estrogen was produced in 1913, and the pure compound was isolated by the Americans Willard Allen and Alan Doisy in 1923. At this time an Austrian physiologist, Ludwig Haberlandt, was carrying out experiments on rabbits to apply the new-found knowledge of hormones for contraceptive ends. By 1927 he was able to write, "It needs no amplification, of all methods available, hormonal sterilization based on biologic principles, if it can be applied unobjectionably in the human, is an ideal method for practical medicine and its future task of birth control." Hostile public attitudes made research on birth control virtually impossible, however, and Haberlandt's ideal was not realized until the 1960s.

Onanism

Users. In the Old Testament story of Onan (Genesis 38:8–10), Judah ordered his son Onan to sleep with Onan's recently widowed sister-in-law, but Onan refused on the ground that "the descendants would not be his own, so whenever he had relations with his brother's wife, he let [the seed] be lost on the ground." As a punishment God killed him, although it is unclear whether the punishment was for his practice of coitus interruptus or for filial disobedience. Perhaps the earliest first-person account of contraception comes from the verbatim records of the Inquisition. During a trial of Albigensian heretics from the village of Montaillou in France in the early 14th century, Beatrice, the mistress of one of the accused, berates her lover, asking "What shall I do if I become pregnant by you?" He replies, "I have a certain herb. If a man wears it when he mingles his body with that of a woman he cannot engender, nor she conceive." The method was almost certainly mystical and inefficacious. James Boswell in his *London Journal, 1762–63* records a more practical experience (for May 10, 1763) when he picked "up a strong young jolly damsel, led her to Westminster Bridge and there, in armour complete, did I enjoy her upon this noble edifice." It is notable that, prior to the Industrial Revolution, most accounts of the use of contraceptives relate to illicit sex.

The 17th-century European upper classes, many of whom had their infants wet-nursed, felt the pressure of excess births within marriage, both physically and emotionally. A French aristocrat writing in 1671 to her daughter, who had borne three children by age 22, recommends, "Continue the nice custom of sleeping separately and restore yourself . . . I kiss your husband. I like him even better in his apartment than in yours." Queen Victoria later expressed a similar sentiment: "Men never think, at least seldom think, what a hard task it is for us women to go through [childbirth] *very often.*"

Decline in birth rates as a result of birth control

In the 19th century better diet, more stable political conditions, and improvements in water supply and hygiene and other simple advances in public health began to bring down the death rate. For the first and probably the last time in the history of industrialized nations a large family became the rule. Eighteenth-century France had seen an overall decline in the birth rate, probably brought about by increasing use of coitus interruptus, and most of western Europe followed suit in the 19th century. In 1860 a quarter of all marriages in England and Wales had eight or more children, but by 1925, 50 percent had only one or two children and one in six was childless. In the United States a similar decline in fertility began slightly later: in 1830 the crude birth rate for white Americans was 50 per 1,000, but by 1930 it was only 18 per 1,000.

Among English couples married before 1910 only 15 percent used a method of birth control, while among those married in the years 1935–39, 66 percent used a method. In 1982 in the United States 67.9 percent of married couples aged 15 to 44 used a contraceptive method and another 14 percent were seeking to be pregnant, were pregnant, or had just delivered. There was little variation by religion or race (61 percent of black couples and 69.6 percent of white couples using a method). The commonest method was female sterilization (one-quarter of all users), followed by the contraceptive pill (one-fifth). About 15 percent of couples used condoms, and another 15 percent relied on male sterilization. Fewer than one in 20 couples used periodic abstinence.

In developing countries where family planning services have been emphasized by the government or private organizations, prevalence of contraception usually rises rapidly. In Thailand, for example, use jumped from 15 percent in 1970 to nearly 60 percent in 1981. In Mexico it rose from 30 percent in 1976 to more than 40 percent in the 1980s and in Bangladesh from 8 percent in 1975 to more than 20 percent in 1984. There has been less success, however, in countries with weak birth control services.

SOCIAL AND POLITICAL ASPECTS OF BIRTH CONTROL

Thomas Malthus

Early advocates. In 1798 Thomas Malthus wrote *An Essay on the Principle of Population.* It posed the conundrum of geometrical population growth's outstripping arithmetic expansion in resources. Malthus, who was an Anglican clergyman, recommended late marriage and sexual abstinence as methods of birth control. A small group of early 19th-century freethinkers, including Jeremy Bentham, Francis Place (himself the father of 15 children), and John Stuart Mill, suggested more pragmatic birth control methods such as coitus interruptus, vaginal barriers, and postcoital douching. Robert Dale Owen, the son of a Scottish social reformer, helped spread these revolutionary ideas in North America, and in 1832 a Massachusetts physician and freethinker, Charles Knowlton, wrote a slim book called *The Fruits of Philosophy: or The Private Companion of Young Married People.* Although Knowlton's first edition was published anonymously, he was fined and imprisoned. The book appeared in England two years later and continued to be read for the next 50 years. In 1876 a Bristol publisher was prosecuted for selling *The Fruits* *of Philosophy.* Incensed, Charles Bradlaugh, the leader of Britain's National Secular Society and subsequently a member of Parliament, and Annie Besant reissued the pamphlet and notified the police. They were charged and tried, the public prosecutor claiming "this is a dirty, filthy book," but the conviction was quashed on grounds of a faulty indictment. The trial received wide publicity and, through the national press, brought birth control onto the breakfast table of the English middle classes at a time when, for economic reasons, they were eager to control their fertility. The Malthusian League, founded some years earlier by George Drysdale, began to attract wide public support. Similar leagues began in France, Germany, and The Netherlands, the latter opening the world's first family planning services, under Dr. Aletta Jacobs, in 1882.

Margaret Sanger and Marie Stopes

But it was two women, Margaret Sanger in the United States and Marie Stopes in Britain, who were to make birth control the object of a national, and ultimately global, social movement. Both used the controversy that surrounded birth control as a ready way of attracting attention. Sanger, a trained nurse, encountered miserable conditions in her work among the poor. She was inspired to take up her crusade when she attended a woman who was dying from a criminally induced abortion. In 1914 she started a magazine, *The Woman Rebel,* to challenge laws restricting the distribution of information on birth control. She was indicted and fled to Europe, but when she returned to stand trial in 1916 the charges against her were dropped. Later that year she opened a family planning clinic in Brownsville, Brooklyn, New York, but the police immediately closed it, and Sanger was arrested and convicted on charges of "maintaining a public nuisance." After many vicissitudes, a compromise was struck and family planning clinics were allowed in the United States on the condition that physicians be involved in prescribing contraceptives. In 1936 a New York court, in a case known as *United States* v. *One Package of Japanese Pessaries,* ruled that contraceptives could be sent through the post if they were to be intelligently employed by conscientious physicians for the purpose of saving life or promoting the well-being of their patients.

The movement for birth control was led in Britain by Marie Stopes, the daughter of a middle-class Edinburgh family. She was one of the first women to obtain a doctorate in botany (from the University of Munich in 1904). In 1918 she published an appeal for sexual equality and fulfillment within marriage, *Married Love,* which at the time was considered to be a radical text. Margaret Sanger met Marie Stopes and persuaded her to add a chapter on birth control. While Sanger's advocacy emphasized the alleviation of poverty and overpopulation, Stopes sought as well to relieve women of the physical strain and risks of excessive childbearing. Her *Married Love* was followed by *Wise Parenthood* (1918), and in 1922 Stopes founded the Society for Constructive Birth Control and Racial Progress.

The population explosion. In 1790 a Venetian monk, Gianmaria Ortis, concluded that human population growth could not continue indefinitely. Malthus' work a few years later stimulated more discussion and also provided the intellectual clue that inspired Charles Darwin's theory of biological evolution through the survival of the fittest. The debate about human numbers remained academic, however, until the 1950s, when a surge in population occurred as a result of the comparative peace and prosperity following World War II.

In Malthus' time world population was under 1,000,000,000, and when Sanger and Stopes opened the first birth control clinics population was still less than 2,000,000,000. In 1960 global population surpassed 3,000,000,000, and the next 1,000,000,000 was added in a mere 15 years. In the 19th century the population of industrialized nations rarely grew by more than 1 percent per annum, but in the 1960s and '70s many developing countries exploded at a rate of 2 to 3 percent per year.

Rapid population growth has several economic consequences. It requires heavier investment in education, health, and transport merely to maintain these services at their previous level; yet, the working population has a higher burden of dependence to support, making both individual and national saving more difficult. Although population growth is not the only problem dividing rich and poor countries, it is one important variable that has widened the gap in growth in per capita income between developed and developing nations. Advocates of birth control see it as a means to prevent the personal and social pressures that result from rapid population growth.

BIRTH CONTROL AND HEALTH

Risk to the mother and infant

There is a marked relationship between patterns of reproduction and the risk of death to the mother and her child. Maternal deaths and infant mortality are up to 60 percent higher among girls under 15 than among women who have a child in their early 20s. The risk of death to the mother and her child rises again in the second half of the 30s. Maternal and infant mortality is lowest for the second and third deliveries. The risk of certain congenital abnormalities, such as Down's syndrome (mongolism), is also greater in older women. Therefore, patterns of sexual abstinence and birth control, which concentrate childbearing in the age group 20–35 and limit family size to two or three children, have a direct impact on public health.

At the same time, it must be recognized that patterns of human reproduction have been finely tuned over millions of years of evolution and the postponement of childbearing until the later 20s or 30s also increases the risk of certain diseases. In particular, cancer of the breast is more common in women who postpone the first birth until the later 20s or older. In the Western world, the risk of death to women in childbirth is approximately one in 10,000, but in developing countries, where half the children born are delivered by traditional birth attendants, it is often 10 times as high. As the number of births worldwide rises, a greater number of women are likely to die having children. Simple access to birth control may be expected to reduce high death rates.

METHODS OF BIRTH CONTROL

Nonmedical methods. *Abstinence.* Abstinence is important in many societies. In the West, most individuals abstain from regular sexual intercourse for many years between puberty and marriage. Raising the age of marriage has been an important element in the decline of the birth rate in China, Korea, and Sri Lanka. Abstinence among couples with grown children is important in some traditional societies, such as certain Hindu groups.

Breast-feeding. The role of breast-feeding in the regulation of human fertility can be illustrated by the following calculation: in Pakistan breast-feeding is virtually universal, and many women breast-feed for two years or more. Fewer than one in 10 women use a modern method of contraception; but if breast-feeding were to decline to levels now found in Central America, four out of 10 women would have to use an artificial method of birth control just to prevent the fertility rate from rising.

Although the information is important to demography, there is no simple way to predict when an individual breast-feeding woman will become fertile again. If she seeks security against pregnancy, a woman may in fact have an overlap of several months between the time she adopts an artificial method and the end of her natural protection.

Coitus interruptus and coitus reservatus

Coital techniques. Coitus interruptus, the practice by which the male withdraws the penis prior to ejaculation, has been an important method of birth control in the West and was used by more than half of all British couples until well after World War II. It is most common among Roman Catholic and Islāmic groups but is less used in the Orient, where coitus reservatus (intercourse without ejaculation) may be more common. The failure rate for coitus interruptus (five to 20 pregnancies per 100 women-years of exposure) overlaps with that of barrier methods of birth control. Although frequently condemned by those promoting other methods of family planning, there is no evidence that coitus interruptus causes any physical or emotional harm. There may be preferable ways of controlling fertility, but for many couples coitus interruptus is better than no method.

The belief that conception cannot take place unless the woman has an orgasm is widespread but untrue. Postcoital douching is not an effective method of birth control.

Barrier methods. Modern high-quality condoms have the advantage of simplicity of use and anonymity of distribution. They are sold in pharmacies, in supermarkets, through the mail, and even in barber shops and at news stands and have been used by more than half of British and American men at one time or another. Use is most extensive in Japan. The acceptance of condoms has been increased in recent decades by advances in packaging and lubrication and, more recently, by the addition of a spermicide. When used carefully, condoms can have a failure rate as low as some intrauterine devices (two to five per 100 women-years of exposure).

Many chemicals act as spermicides; one of the most widely used is a detergent, nonoxynol-9, found in most foams, pessaries, and dissolving vaginal tablets. Spermicides are either used alone, when they have a moderate failure rate, or in combination with a barrier method such as a diaphragm or a disposable sponge.

Periodic abstinence. Although a couple may make a private choice to use periodic abstinence, just as they might buy condoms, most modern methods of periodic abstinence require careful training by a trained counsellor. Awareness of human fertility can be valuable when a couple is attempting to conceive a child. The method makes considerable demands on the partners, but if well taught it may also enhance the marital relationship.

Several types of periodic abstinence, also known as the rhythm method or natural family planning, are practiced. The time of ovulation can be estimated from a calendar record of previous menstruation, but this method has low effectiveness. More reliable methods include keeping a daily record of body temperature or recording physical changes in the cervix (the neck of the womb) and cervical mucus (the mucous method, also called the Billings method). These methods may also be combined (sympothermic method). As with several methods of birth control, a wide range of failure rates has been recorded for the various types of periodic abstinence, extending from one pregnancy per 100 women-years of exposure to more than 20 per 100.

Medical methods. *Hormonal contraceptives.* Hormonal contraceptives use artificially synthesized derivatives of the natural steroid hormones estrogen and progesterone. Estrogen is responsible for the growth of the lining of the womb (endometrium), which occurs early in the menstrual cycle. Progesterone is produced in the second half of the cycle and in great quantities in pregnancy. It makes the mucus in the lower part of the reproductive tract resistant to the ascent of sperm and also alters the lining of the womb. Both hormones cause changes in the breasts and elsewhere in the body. They act on the base of the brain and the associated pituitary gland. This gland, in turn, secretes hormones (gonadotrophins) that regulate the production of estrogens and progesterone by the ovaries.

Oral contraceptives

Most oral contraceptives contain a combination of estrogen and progesterone. The combination, like the hormone balance of normal pregnancy, prevents the release of eggs from the ovaries. A minority of pills contain only a progestogen (a progestational steroid) and act mainly by causing changes in the mucus that prevent the ascent of sperm. In different doses, combination pills and certain other hormonal preparations can be used after coitus. They prevent pregnancy up to two or three days after the fertilizing intercourse, primarily by rendering the lining of the womb unsuitable for the attachment (implantation) of a fertilized egg.

More than 100,000,000 women currently use oral contraceptives or have used them in the past. In many countries pills are widely distributed by community workers and through pharmacies, without direct medical supervision. Injectable contraceptives are registered for use in more than 80 countries, including most of the Third World, the United Kingdom, Sweden, and New Zealand. The injectable preparation Depo-Provera has had a particularly controversial history, having been referred for further study by the U.S. Food and Drug Administration in 1974, 1978, and 1984. Research has been undertaken on subdermal implants and intravaginal rings (which slowly release hormones for absorption through the vaginal wall). In China a once-a-month pill is available.

Hormonal contraceptives belong to the 20th century. Slow to be developed, sometimes misunderstood by physicians, and often the centre of the news media's attention, they have been alternately oversold and overcriticized. Nevertheless they have wrought a medical and social revolution. They are remarkably effective, cheap to manufacture, and relatively simple to use. But as methods that imitate, albeit imperfectly, the menstrual cycle and some of the changes normally occurring in pregnancy, they are responsible for a wide range of good and bad changes in the body.

Risks and benefits of oral contraceptives

As noted, the principle of hormonal contraception was understood in the 1920s, but it was 30 years before the drive of Margaret Sanger (then more than 70 years old) and the philanthropy of Mrs. Page McCormick were to draw the first oral contraceptive preparations from somewhat reluctant scientists and physicians. The first clinical report of the use of oral steroid hormones to suppress ovulation was published by Gregory Pincus and John Rock from Boston in 1956. The approval of the U.S. Food and Drug Administration was granted in 1960, and marketing of the preparations in Britain began two years later. When oral contraceptives are used correctly, fewer than one woman in 100 per year of use will conceive an unintended pregnancy. A woman's menstrual cycle is more regular when she uses the pill, and users tend to be less anemic than nonusers. Immediate adverse side effects can include nausea, breast tenderness, headaches, and weight gain. But it was only after the first few million women had used the method for some years that important but rare side effects began to be reliably documented and accurately measured. Predictably adverse conditions leading to death or serious disease were discovered before a number of beneficial, and even lifesaving, effects were demonstrated. The order of these discoveries, together with the perceived social impact of the method, probably accounts for much of the controversy that has surrounded and continues to surround oral contraceptives.

Large-scale epidemiological research involving tens of thousands of women has now demonstrated that users of the pill are more likely than nonusers to suffer from heart attacks, strokes, and blood clots in the veins. These effects are extremely rare in younger women, but occurrence is multiplied several times in all age groups among users who smoke. Users of oral contraceptives are less likely than nonusers to develop cancer of the ovary or uterus. Use reduces the chance of benign breast disease but neither protects against nor causes breast cancer. The risk of pelvic infection is approximately halved among users. Fertility returns rapidly after discontinuing use, and, while some artificial steroids in high doses can damage the fetus, there is no consistent evidence that oral contraceptives cause congenital abnormalities.

It is difficult to balance the list of the oral contraceptive's risks and benefits, some of which (such as the small risk of heart disease) appear when use begins while others (such as protection against certain forms of cancer) only develop after several years of use but persist even after use has stopped. Overall, taking all known risks and benefits into account, the average woman in a Western nation actually increases her life expectancy by a small but calculable amount if she uses oral contraceptives, while the older woman, especially if she smokes, is at a small but measurably higher risk of death. In Western nations women over 40 and those over 35 who smoke are usually advised to use another method of birth control. Among women in Third World countries the risks of death from childbirth remain many times greater, and, although the pill has not been as closely studied in such settings, the advantages of its use are almost certainly correspondingly greater.

Male contraception. In normal circumstances a man can produce several million sperm per day and is almost continually fertile. A woman's menstrual cycle, with predictable time of ovulation, is medically much simpler to control. Research on a male pill has been disappointing. Sperm production has been controlled under experimental conditions, and in China a substance called gossypol, derived from the cottonseed, has been used as an oral contraceptive for males. Most substances used in the control of male fertility, however, either have proved toxic or have depressed sexual drive as well as sperm count.

Intrauterine devices. Almost any foreign body placed in the uterus will prevent pregnancy. While intrauterine devices (IUD's) were invented in the 19th century, they only came into widespread use in the late 1950s, when flexible plastic devices were developed by Jack Lippes and others. The IUD, made in a variety of shapes, is placed in the uterus by passing it through the cervix under sterile conditions. Like oral contraceptives, IUD's probably act in several complementary ways. When the IUD is in place an abnormally high number of white blood cells pass into the uterine cavity, and the egg, even if fertilized, is destroyed by the white blood cells before implantation. Nevertheless, one to three out of every 100 users per year will get pregnant with the IUD in place.

An intrauterine device can be inserted on any day of the menstrual period and immediately after a birth or abortion. The advantage of an IUD lies in its long-term protection and relative ease of use. The disadvantages include heavier menstrual flow and an increased risk of uterine infection. Approximately 60 million women use IUD's worldwide. The largest use is recorded in China. IUD's are most satisfactory when used by older women who have had children and are recommended less frequently for young women, primarily because of the risk of pelvic infection.

In 1970 Jaime Zipper, a physician from Chile, added copper to plastic devices, thereby permitting designs that caused less bleeding and increased effectiveness. IUD's that slowly release progesterone derivatives have also been developed.

Voluntary sterilization. More than 100,000,000 couples worldwide have selected sterilization, and the method prevents more pregnancies each year than any other method of birth control. Voluntary sterilization has proved popular in both rich and poor countries, and the number of operations performed is likely to continue to rise. Wherever sterilization of the female (tubal sterilization) has been offered it has proved popular. Fewer male sterilizations (vasectomies) than female sterilizations have been performed worldwide but demand grows consistently wherever a reliable service is offered.

Tubal sterilization and vasectomy

Vasectomy is a quick, simple operation normally carried out under local anesthesia. The vas deferens, the tube carrying the sperm from the testicles to the penis, is blocked, and a number of ejaculations must be made after the operation to remove all the sperm capable of fertilization. Local bleeding and infection can occur after the operation, but no long-term adverse effects have been demonstrated in men. In some animals, however, disease of the blood vessels has been reported to be more common after experimental vasectomy.

The fallopian tubes, which carry the egg from the ovary to the uterus, lie buried deep in the female pelvis. To perform sterilization a surgeon must either open the abdomen, in a procedure called laparotomy, and close the tubes under direct vision, or insert an optical instrument (laparoscope) to view the tubes so that a clip, ring, or electrocautery can be applied. The only proved side effects of female sterilization are those associated with any surgery and local or general anesthesia.

An individual seeking sterilization must accept the operation as irreversible while at the same time understanding that in rare cases, in either sex, the operation can fail even when properly carried out. In cases of extreme need, reversal of both female and male sterilization has been attempted, with more than 50 percent of patients later conceiving children. Surgical reversal is easier for male sterilization.

Abortion. Abortion is the termination of pregnancy less than 28 weeks after the last menstrual period. Until the eighth week of pregnancy the conceptus is called an embryo, and after that time a fetus. Abortion may be spontaneous (miscarriage) or induced, and induced abortions are legal in some circumstances in some countries and illegal in others. An incomplete abortion is one after which part of the conceptus remains in the uterus. It is associated with bleeding and the risk of infection.

Human reproduction is an imperfect process. Only one sperm is necessary for fertilization, yet the male's ejaculate contains millions of sperm. As many as half of the eggs fertilized die within 10 days of fertilization without the woman even knowing she has conceived. As many as one-fifth of recognized pregnancies miscarry. Much of this massive wastage is associated with chromosomal and other abnormalities in the embryo.

Induced abortion

Induced abortion has occurred throughout history and is known in almost all contemporary societies. A variety of herbs and potions have been used over the ages, and physical violence as a cause of abortion is mentioned in the Bible (Exodus 21:22). In the contemporary world tens of millions of abortions are performed annually. Some are deemed legal—*i.e.,* carried out by qualified persons with proper supervision—and others illegal. Massage abortion is common in Southeast Asia. It is usually conducted by a traditional birth attendant who pounds the pregnant abdomen until uterine bleeding commences or pain stops the procedure. In the rest of the world a common method is to pass an object through the neck of the womb to dislodge the placenta. Abortions performed by unqualified persons can endanger the woman's life. In Latin America, for example, approximately 1,000,000 women a year are admitted to hospitals suffering from incomplete abortions, mostly the result of illegal abortion.

Abortion and contraception have a complex relationship during the process of demographic change. A decline in the birth rate may reflect a rise in the number of abortions and the use of contraception. As the rate declines further, abortions peak (as in Japan in the 1950s and 1960s), but, if contraceptive services are readily available, then the number of abortions falls as the number of conceptions falls. If, however, contraceptive services are not readily available (as in the Soviet Union), then the number of abortions remains high.

The commonest technique of inducing legal abortion is vacuum aspiration of the uterine cavity. When completed before the 12th week of pregnancy the procedure is brief and can be done without general anesthesia. It has proved to be remarkably safe for the woman, with a death rate of less than one in 100,000 operations. Scraping (curettage) of the uterus is an older surgical procedure. It is less satisfactory than vacuum aspiration early in pregnancy but can be more easily used after 12 weeks. Late abortions can also be performed by chemical means (the introduction of prostaglandins) or by the injection of urea or salt into the space around the embryo.

FAMILY PLANNING SERVICES

National family planning movements have emphasized the right of the individual to determine family size as well as the contribution family planning can make to national and global population problems. Some methods of birth control, such as coitus interruptus and, in extreme cases, abortion, may involve no person other than the individual or couple. But most methods require manufacture, distribution, promotion, counselling, and in some cases financial subsidy.

The retail trade in contraceptives has been a major element in the spread of contraception and remains important in the developing world. In particular, social marketing programs, which adjust prices to people's needs, have allowed governments to make contraceptives available to large numbers of people quickly and at affordable cost. Private doctors may advise patients about the use of birth control on a confidential basis and may charge a fee.

National and international movements

The first altruistic attempts to offer direct family planning services began with private, pioneering groups and often aroused strong opposition. The work of Sanger and Stopes reached only a small fraction of the millions of couples who in the 1920s and '30s lived in a world irrevocably altered by World War I, crushed by economic depression, and striving for the then lowest birth rates in history. In 1921 Sanger founded the American Birth Control League, which in 1942 became the Planned Parenthood Federation of America. In Britain the Society for the Provision of

Birth Control Clinics was to evolve into the Family Planning Association. As early as 1881 the British Malthusian League had brought together individuals from 40 nations to discuss birth control, and five genuinely international meetings had taken place by 1930. A conference was held in Sweden in 1946. The first birth control clinic in India opened in 1930, and in 1952 in Bombay, Margaret Sanger took the first steps toward creating what became the International Planned Parenthood Federation (IPPF).

The modern era in international family planning opened in the second half of the 1960s when governments, beginning with Sweden, gave money to support the worldwide work of the IPPF. William Draper lobbied with particular effectiveness in the United States to build up the IPPF and to put together the United Nations Fund for Population Activities (UNFPA), established in 1969. For several years the U.S. Agency for International Development helped to support the IPPF and the UNFPA. The United Nations held international conferences on population in Bucharest in 1974 and Mexico City in 1984.

THE LEGALITY OF BIRTH CONTROL

The Comstock Act

In the 19th century the law was used as an assertion of existing morality. In the United States Anthony Comstock lobbied to pass an Act for the Suppression of Trade in, and the Circulation of, Obscene Literature and Articles of Immoral Use. When asked why he classified contraception with pornography, Comstock answered, "If you open the door to anything, the filth will pour in." Anti-contraceptive and anti-sterilization clauses were added to the Napoleonic Code applying to France and French colonies. In Britain, however, the law never specifically condemned contraception or sterilization, and Bradlaugh and Besant were accused under the Obscene Publications Act.

The 20th century has seen statute laws used as a vehicle of social change and as a battleground of conflicting philosophies. The Nazi Third Reich invaded the bedrooms of its citizens before it moved its troops into the Sudetenland and Czechoslovakia. It forbade the display of contraceptives, which it condemned as the "by-product of the asphalt civilization." By contrast, the Proclamation of Teheran in 1968 (paragraph 16) provided "Parents have a basic human right to determine freely and responsibly the number and spacing of their children." This concept was written into Yugoslavia's constitution, and China officially made family planning an obligation for each citizen. U.S. courts interpreted the constitutional right of privacy to include birth control choices when the Comstock Act was finally overthrown in the cases of *Griswold* v. *Connecticut* (1965) and *Eisenstadt* v. *Baird* (1972). In Ireland the case of Mary McGee (1973) reversed an Irish anti-contraceptive law of 1935, and in the Luigi deMarchi case in 1971 the Italian Supreme Court struck down the Fascist laws limiting the availability of contraception. At the other extreme, Singapore has passed legislation removing certain tax credits from couples with three or more children.

Laws prohibiting abortion

By the end of the 19th century almost every nation in the world had passed antiabortion legislation. In the United States restrictive laws were propelled not so much by moral considerations as by the desire of the medical profession to regulate the practices of unqualified doctors.

The 20th century has seen the pendulum swing in the opposite direction, and 70 percent of the world's population now lives in countries where abortion is legally available. The Soviet Union (1920) became the first country in the 20th century to permit legal abortion, and the Scandinavian and most Eastern European countries had liberal abortion laws by the late 1960s. In Britain the Offenses Against the Person Act of 1861 was reversed by the 1967 Abortion Law, and by 1970 Canada and several U.S. states (including New York State) had passed abortion reform legislation. Arguments usually centred on hard cases, such as that of a woman carrying an abnormal fetus or living in extreme poverty. On January 22, 1973, the U.S. Supreme Court struck down as unconstitutional all antiabortion laws remaining in the United States. The Court argued "that the right of personal privacy includes the abortion decision." India, China, Australia, Italy, France, The Netherlands, and many other countries decided to permit abortion under statute law or following individual case precedents. It has always been difficult to harmonize statute law with biological processes, and several new therapies, such as the use of drugs to induce delayed menstruation, and even the use of IUD's, have not been clearly defined as falling under the category of either contraception legislation or abortion legislation.

In this most controversial aspect of birth control, legal positions have oscillated, depending on circumstance and on government. In 1935 Joseph Stalin reversed Lenin's liberal abortion law in the Soviet Union, and the Nazis declared abortions to be "acts of sabotage against Germany's racial future." In 1942 a woman was guillotined in Nazi-dominated France as a punishment for abortion, and in 1943 the government of the Third Reich introduced the death penalty for abortionists who "continually impaired the vitality of the German people." After the defeat of the U.S. antiabortion laws in 1973, a strong drive was undertaken by antiabortionists in the United States to limit the interpretation of the Supreme Court ruling and, if possible, to reverse that ruling by congressional action, constitutional amendment, or the appointment to the Supreme Court of justices who were against abortion.

The law, by defining marriage age, regulating medical practice, and controlling advertising and such factors as the employment of women, also affects many other variables that determine the size of a family. For example, Section 4(5) of the 1954 British Television Act prohibits the advertising of matrimonial agencies, fortune-tellers, and contraceptives.

ETHICS AND THE INFLUENCE OF RELIGIOUS SYSTEMS

The ethics of birth control has always been a topic of debate. All of the world's major religions endorse responsible parenthood, but when it comes to methods the consensus often dissolves. Hindu and Buddhist teachings are linked by a belief in reincarnation, but this has not been extended to an obligation to achieve maximum fertility. The Buddhist religion requires abstinence from any form of killing, and strict Buddhist groups have interpreted this requirement as support for opposition to contraception. At the same time, Buddhist scripture contains the phrase "Many children make you poor," and the few prevailing constraints against birth control have been interpreted as affecting individuals, not state policy.

In the Muslim religion, the Prophet Muḥammad endorsed the use of *al-azl* (coitus interruptus) for socioeconomic reasons and to safeguard the health of women. The Qu'rān instructs, "Mothers shall give suck to their offspring for two whole years if they desire to complete their term" (II,233). In general, modern methods of family planning have been accepted by Islāmic religious leaders, although sterilization is resisted as mutilation. Some fundamentalist Islāmic groups, most notably in Iran in the 1980s, have opposed family planning in general.

The Judeo-Christian tradition has been more divided in its approach to birth control; and Europe and North America have had a disproportionate role in medical research and practice. Until the Industrial Revolution in the West, artificial methods of birth control seemed irrelevant or even antagonistic to reproduction and to the spiritual goals of marriage. Christendom was very slow to recognize new medical knowledge and new social needs, thereby retarding the development of birth control methods and diffusion of services. For example, in part because of religious objections, the U.S. National Institutes of Health were explicitly barred from research on contraception until 1961.

Historically, Jewish doctrines on marriage and procreation were related to the national struggle for survival and the traditions of a close-knit monotheistic community in which the individual was perpetuated through family. Judaism imposes an obligation to have children, although love and companionship are deemed an equally important goal of marriage. Orthodox sections of Judaism permit women to use certain methods of birth control, especially when necessary to protect the mother's health. Reformed and Conservative branches urge proper education in all methods of birth control as enhancing the spiritual life of the couple and the welfare of mankind. Many Jewish physicians and leaders, such as Alan Guttmacher, have joined in the advocating of birth control.

Augustine's views

The early Christian Church reacted against the hedonism of the later Roman Empire and, believing that the Second Coming of Christ preempted the need for procreation, held celibacy superior to marriage. Early Christians opposed the Gnostic movement that viewed the world as the creation of evil and procreation as the perpetuation of that evil. Instead they supported the Stoic argument that sexual passions distracted man from the contemplation of the One, the True, the Good, and the Beautiful. It was a short step for the 2nd-century theologian Clement of Alexandria to associate sexual intercourse with guilt and argue that it could only be justified by the obvious need to reproduce. Clement even argued that the human soul fled the body during a sexual climax. Augustine (AD 354–430), in his writings, especially in *Marriage and Concupiscence* (AD 418), laid the intellectual foundation for more than 1,000 years of Christian teaching on birth control. He concluded that the male semen both contained the new life and transmitted Adam's original sin from generation to generation.

Among the practices Augustine condemned were not only coitus interruptus (onanism) but also what today would be called natural family planning. Not surprisingly the explicit justification of periodic abstinence by the modern church continues to come into conflict with remnants of Augustine's more pessimistic identification of sex with sin.

An important challenge to traditional Roman Catholic teaching arose in 1853 when the church's Sacred Penitentiary ruled that couples using periodic abstinence were "not to be disturbed." Among all Christian denominations, however, change was halting. In 1920 the Anglican Lambeth Conference condemned "any deliberate cultivation of sexual activity as an end in itself," although by 1930 the Conference had taken some steps toward the moral justification of birth control. By 1958 its members concluded that "implicit within the bond of husband and wife is the relationship of love with its sacramental expression in physical union."

The Roman Catholic viewpoint developed even more slowly. The conservative theologian Arthur Vermeersch drafted much of Pope Pius XI's encyclical *Casti Connubii* (1930), condemning all methods of birth control except periodic abstinence as "grave sin." This teaching was reaffirmed by Pius XII in 1951. The Second Vatican Council (1962–65), however, described marriage as a "community of love" and the council's Constitution on the Church and the Modern World (*Gaudium et Spes*) exhorts parents to "thoughtfully take into account both their own welfare and that of their children, those already born and those which may be foreseen." Once the dual purposes of sexual relations to procreate and to express love had been accepted by the Second Vatican Council, however, some theologians and a great many Roman Catholic couples examined their own consciences and found it increasingly difficult to distinguish between intercourse during intervals of infertility brought about by the use of hormonal contraceptives and intercourse during the infertile intervals of the menstrual cycle. John Rock, who helped to develop the contraceptive pill and was himself a Roman Catholic, argued for just such a reassessment in his book *The Time Has Come* (1963). Gathering pressure led to the establishment of the Commission for the Study of Population and Family Life. It submitted its report to Pope Paul in 1966. Among the commission's members, the medical experts recommended by a vote of 60 to four, and the cardinals by nine to six, to liberalize Roman Catholic teaching on birth control.

The *Humanae Vitae*

In 1968, however, Pope Paul restated the traditional teaching of *Casti Connubii* in his landmark encyclical *Humanae Vitae*, using papal authority to assert that "every conjugal act [has] to be open to the transmission of life." *Humanae Vitae* came as a surprise to most church leaders and left many of the laity in a painful conflict between obedience and conscience. Six hundred Roman Catholic scholars signed a statement challenging *Humanae Vitae*, many episcopates attempted to soften the harsher aspects of the encyclical, a flood of priests left the church, and the number of U.S. Catholics attending mass weekly fell from 70 percent before the issuing of the encyclical to 44 percent a few years afterward. The total marital fertility (the number of children in a completed family) of U.S. Catholics (2.27 in 1975) became virtually the same as that of non-Catholics (2.17). At the same time a new movement began within the Roman Catholic Church, taking strength and inspiration from *Humanae Vitae*. Among lay organizations, the International Federation for Family Life Promotion was founded in 1974 and the Family of the Americas Foundation (formerly World Organization of the Ovulation Method—Billings; WOOMB) was founded in 1977.

The Eastern Orthodox Church maintains that parenthood is a duty. While it considers the use of contraception to be a failure in spiritual focus, the church has not sought to hinder the distribution of birth control information or services.

Misuse of birth control

Birth control, like other technologies, can be misused. In the 19th century vasectomy was used for men judged to be compulsive masturbators, and a century later, during the state of emergency declared in India in 1975, the Indian government supported forcible sterilization of low-caste men as part of a population control program. In the not too distant past unmarried women in the Western world who became pregnant faced such hostility from society in general that the majority felt they had no choice but illegal abortion, while in China today women are subject to intense social pressure to legally abort a second or subsequent pregnancy inside marriage. In contemporary Western society conventional restraints on sexual experience prior to marriage are in turmoil. Vigorous debate centres on the question of whether the availability of birth control to young people encourages premarital sexual relations or avoids unplanned pregnancies that otherwise might occur. Certainly, similar patterns of availability of contraceptives may be observed in markedly different social settings with high and low incidence of premarital sex (for example, the United States and China, respectively). There is no evidence that the availability of birth control either encourages or discourages particular patterns of sexual behaviour.

Personal opinions on abortion

The debate over the ethics of induced abortion can arouse deep divisions even in otherwise homogeneous groups. At one extreme abortion is considered to be the moral equivalent of murder and the life of the fetus is held to take precedence over that of the pregnant woman. At the other extreme it is argued that a woman has an absolute right over the pregnancy within her body. Surveys of opinion show that most people find abortion to be a sad and complex topic. The majority would prefer not to experience abortion but nevertheless feels that abortion is justified in certain cases, such as when tests show evidence of congenital abnormality, when pregnancy results from sexual crimes, or when the parents live in extreme poverty. The embryological discoveries of the past century cannot solve the metaphysical questions posed in the past. The U.S. Supreme Court decision on abortion in 1973 concluded "We need not resolve the difficult question of when life begins. When those trained in the respective disciplines of medicine, philosophy, and theology are unable to arrive at any consensus, the judiciary, at this point in the development of man's knowledge, is not in a position to speculate as to the answer." In short, the definition individuals give to the beginning of life determines their judgment about the acceptability or licitness of abortion, and those definitions remain in the sphere of wholly human judgment.

CONCLUSIONS

Modern mankind can never return to the way of life that characterized most of human evolution. Settled agriculture and, to an even greater extent, urban living have irrevocably altered natural, finely tuned patterns of human reproduction. New social and artificial restraints on fertility must replace high infant mortalities and the invisible but important physiological controls that once limited family size. The variables that encourage small families are still not fully understood, but they include urbanization, educational and employment opportunities for women, and easy access to family planning services. In a traditional agricultural society children bring hope of economic rewards to their parents at an early stage by sharing in the work that is necessary to support the family, whereas in modern industrial societies the care and educating of children represent long years of heavy expenditure by the parents. This switch in the cost of children may be the most important factor determining the adoption of family planning.

Western societies took more than a century to reach zero population growth and adjust to the rapid expansion of population that accompanied their industrialization. Most of the changes that occurred in patterns of family planning took place before public family services were established and at considerable emotional and physical costs to many couples. By contrast, the majority of the governments of contemporary Third World countries have established national family planning policies and actively encourage the use of public family services. The World Fertility Survey shows that more couples in developing countries desire small families than actually achieve their goals.

The significance of the choices facing policymakers and individual families can be illustrated by reference to trends in family planning in the People's Republic of China. For a generation after the Revolution of 1949 national leaders maintained that a Communist economy could accommodate any rate of population growth, and family planning services, while available, were not emphasized. As a result of the rapid population growth in the 1950s and '60s, however, the number of marriages in China exceeds by 10,000,000 each year the number of fertile partnerships broken by death or by the onset of the woman's menopause. In order to try to stabilize the population, the Chinese government has recommended that 50 percent of rural couples and 80 percent of urban couples have only one child. The application of this type of policy can have an ironic effect on individual women: older women belong to a generation that could not always obtain birth control services, and younger women may be encouraged or, in some cases, even forced to abort pregnancies they would like to keep.

Although consensus has not been reached on the range of birth control methods society should offer to individual members, the right of couples to determine the number and spacing of their children is almost universally endorsed, while the possibility of coercive family planning is almost as widely condemned. Throughout the world, awareness of the advantages and disadvantages of specific methods of birth control, thoughtful judgments about ethics, and further evolution in medical and scientific knowledge will continue to be important to the welfare of the family, of individual nations, and of the entire globe.

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(M.P.)

Bismarck

Otto von Bismarck—or Otto Eduard Leopold, Prince von Bismarck, Count von Bismarck-Schönhausen, Duke von Lauenburg—was a Prussian statesman who in 1871 founded the German Empire and served as its first chancellor for 19 years. Once the empire was established, he actively and skillfully pursued pacific policies in foreign affairs, succeeding in preserving the peace in Europe for about two decades. But in domestic policies his patrimony was less benign, for he failed to rise above the authoritarian proclivities of the landed squirearchy to which he was born.

Archiv fur Kunst und Geschichte

Bismarck, 1880.

LIFE

Early years. Bismarck was born on April 1, 1815, at Schönhausen, near Magdeburg, in the northeastern provinces of Prussia. His father, Ferdinand von Bismarck-Schönhausen, was a Junker squire descended from a Swabian family that had ultimately settled as estate owners in Pomerania. Ferdinand was a typical member of the Prussian landowning elite. The family's economic circumstances were modest—Ferdinand's farming skills being perhaps less than average—and Bismarck was not to know real wealth until the rewards flowed in after the achievement of German unification. His mother, Whilhelmine Mencken, came from an educated bourgeois family that had produced a number of higher civil servants and academics. She had been married to Ferdinand von Bismarck at age 16 and found provincial life confining. When her son Otto was seven, she enrolled him in the progressive Plamann Institute in Berlin and moved to the capital to be near him. The young Bismarck resented exchanging an easy life in the country for a more circumscribed life in a large city, where in school he was pitted against the sons of Berlin's best-educated families. He spent five years at the school and went on to the Frederick William gymnasium for three years. He took his university entrance examination (*Abitur*) in 1832.

Education

With his mother's encouragement, he took up the study of law at the University of Göttingen in the kingdom of Hanover. Evidently Bismarck was a mediocre student who spent much of his time drinking with his comrades in an aristocratic fraternity. After a brief stint at the university in Berlin, he entered the Prussian civil service, where he was plagued by boredom and an inability to adhere to the hierarchical principles of the bureaucracy. His mother's death in 1839 gave him the opportunity of resigning in order to come to the assistance of his father, who was experiencing financial difficulties in the management of his estate. From 1839 to 1847 Bismarck lived the ordinary life of a Prussian country squire. Subsequently he was to romanticize these years on the land and wonder why he had abandoned an idyllic existence for the insecurities of a life in politics. This frequently expressed nostalgia may have been more guise than reality.

During this period he met and married Johanna von Puttkamer, the daughter of a conservative aristocratic family famed for its devout pietism. While courting Johanna, Bismarck experienced a religious conversion that was to give him inner strength and security. A subsequent critic was to remark that Bismarck believed in a God who invariably agreed with him on all issues. There is no question that the marriage was a very happy one. Bismarck's last words before dying in 1898 expressed the wish that he would see Johanna, who had died some years earlier, once again.

His politics during the 1840s did not diverge substantially from those of a typical country squire; if anything, they were more conservative. He believed in a Christian state that received its sanction ultimately from the deity. The existing social and political order was to be defended in order to prevent a Hobbesian chaos of all against all. Given his views, Bismarck was welcomed as a member of the religious conservative circle around the brothers von Gerlach, who were stout defenders of the noble estate against the encroachments of bureaucratic centralization. Bismarck had nothing but sarcasm for aristocratic liberals who viewed England as a model for Prussia. In 1847 he attended the Prussian United Diet, where his speeches against Jewish emancipation and contemporary liberalism gained him the reputation of a backwoods conservative, out of touch with the dynamic forces of his age.

Bismarck's response to the liberal revolution that swept through Europe in 1848 confirmed his image as a reactionary. He opposed any concessions to the liberals and expressed contempt for the king's willingness to bargain with the revolutionaries. He even considered marching his peasants to Berlin to free Frederick William IV from the baneful influence of the rebels. With other archconservatives, including Ernst Ludwig von Gerlach, he began contributing to the *Kreuzzeitung* newspaper (1848) as an organ of antirevolutionary sentiment.

Response to the revolution of 1848

For Bismarck's future role, it is important to understand his analysis of the revolution. Like his contemporary Karl Marx, he identified the forces of change as confined solely to the educated and propertied middle class. The vast majority of Prussians, however, were peasants and artisans, who, in Bismarck's view, were loyal monarchists. The task of the forces of order was to confirm the loyalty of these two groups by means of material concessions. The economic policies of the urban middle-class radicals were rooted in pure self-interest, he maintained. The radicals would spur industrial growth at the expense of the lower middle class and the farm population. Ultimately, even the middle class itself might be won over by tactical concessions and success in foreign policy. This strategic and opportunist thinking distanced Bismarck from the ideological conservatives, who were wedded to traditional concepts of authority. His vision of a manipulative state that sustained its power by rewarding obedient groups remained with him throughout his political career.

Early career. In 1849 he was elected to the Prussian Chamber of Deputies (the lower chamber of the Prussian Diet) and moved his family to Berlin. At this stage he was far from a German nationalist. He told one of his fellow conservatives, "We are Prussians, and Prussians we shall remain. . . . We do not wish to see the Kingdom of Prussia obliterated in the putrid brew of cosy south German sentimentality." In 1851 Frederick William IV appointed Bismarck as the Prussian representative to the federal Diet in Frankfurt, a clear reward for his loyalty to the monarchy.

With the defeat of the revolution in central Europe, Austria had reasserted its supremacy in the German Confederation, and Bismarck, being an archconservative, was assumed to support the status quo, which included Austrian hegemony. He lived in Frankfurt for eight years, where he experienced a commercial and cultural environment quite different from that of a Prussian estate.

It was in Frankfurt that Bismarck began to reassess his view of German nationalism and the goals of Prussian foreign policy. Not only did he find the constant deference to the Austrians in Frankfurt demeaning, but he also realized that the status quo meant acceptance of Prussia as a second-rate power in central Europe. In 1854 he opposed close cooperation with Austria, arguing that it entailed "binding our spruce and seaworthy frigate to the wormy old warship of Austria." Gradually he began to consider the options that would make Prussia the undisputed power in Germany. A vision of a Prussian-dominated northern Europe and a redirection of Austrian power to the Slavic areas in the south took shape in his mind. If necessary, a war with Austria to destroy its hegemony was not to be excluded. Implementation of such a policy would be anything but conservative because it would entail radical changes in the map of Europe as it had been drawn by the conservative powers at Versailles in 1815.

Experience in foreign affairs

Prime minister. In 1859 Bismarck was sent to Russia as Prussian ambassador, and not long thereafter (May 1862) he moved to Paris as ambassador to the court of Napoleon III. Thus he had 11 years of experience in foreign affairs before he became prime minister and foreign minster of Prussia in September 1862. He had come to know personally the architects of French, Russian, and Austrian foreign policy. Ironically, Bismarck was called back by King William I (1861–88) to the reigns of power at a critical juncture in Prussia's internal development.

For more than two years William had been locked in a battle with the Chamber of Deputies over military reform. Having been in the army much of his adult life, the monarch (similar to earlier Prussian kings) considered it entirely within his prerogative to increase the size of the military and the years of service. When the liberal majority did not approve the revenue for these reforms, William refused to negotiate or compromise with liberal politicians over the fundamental issue of sovereignty. He prorogued Parliament twice, and each time the liberal majority increased.

The appointment of Bismarck was the monarch's last desperate effort to avoid parliamentary sovereignty over the military. The Chamber of Deputies interpreted it as an act of defiance—a throwing down of the gauntlet. But the Bismarck who returned to Berlin from Paris was not the backwoods conservative of 1848. Having lived in Frankfurt and Paris, he had come to appreciate the growing importance of the propertied and educated middle class. And in France he had experienced the Bonapartist regime of Napoleon III, which relied on the combination of success in foreign policy and plebiscites at home to shore up the emperor's authoritarian regime. Bismarck had changed to such a degree that he actually returned with the idea of seeking a compromise over the military issue. But William I rejected a sensible proposal offered by Bismarck, leaving him no alternative but a policy of confrontation. Bismarck then announced that there was a "hole" in the constitution. If the king and the members of the Upper Chamber and the Chamber of Deputies, who together were responsible for the budget, failed to come to an agreement, the government in the interim had to proceed without it. Taxes were to be collected (and spent) on the basis of the old budget because civil servants had to be paid and the government had to continue functioning. This tactic, applied from 1863 to 1866, allowed him to implement the military reforms without the sanction of Parliament. Bismarck did, indeed, appear to be the reactionary, confrontational aristocrat out of tune with his time.

But there were hints that this was more appearance than reality. Bismarck said that "Prussia must collect and keep its strength for the right moment, which has been missed several times already; Prussia's frontiers as laid down by the Vienna treaties are not conducive to a healthy national life; it is not by means of speeches and majority resolutions that the great issues of the day will be decided—that was the great mistake of 1848 and 1849—but by blood and iron." He was giving the opposition evidence that he intended to use Prussia's military might not for internal suppression but for the liberal goal of achieving national unification. The liberal opposition, however, chose to ignore these hints, and on May 22, 1863, by a vote of 239 to 61, they informed William I that they would not deal with his prime minister any further. After eight months in office, Bismarck had failed to achieve any agreement with the parliamentary opposition.

Bismarck now turned to foreign policy in the hope that success on this front would weaken the electorate's clear desire for political reform. Trouble had been brewing since 1848 between the Danes and the German population of the duchies of Schleswig and Holstein. (Both duchies were in union with Denmark; Schleswig, however, had a large German population, and Holstein was a member of the German Confederation.) When the Danish king acted rashly, Bismarck made sure that it was Prussia and Austria rather than the German Confederation which represented German interests. Liberal leaders like Rudolf Virchow still saw Bismarck as an unrepentant reactionary who was "no longer the man who joined us with feeling that he was going to accomplish something with an energetic foreign policy."

Convention of Gastein

A quick successful war against Denmark left the fate of Schleswig and Holstein up to Bismarck and the Austrians. After much haggling, the Convention of Gastein was signed on Aug. 20, 1865; it provided for Schleswig to be administered by Prussia and Holstein by Austria. Liberals remained unappeased by Prussian military prowess and once again defeated the army bill in January 1865.

In 1866 Bismarck nonetheless continued his efforts to divert liberal interest from the budget conflict and toward the success of Prussian arms. He repeatedly told the Austrians that their future lay in the south and that they would be wise to yield dominance in Germany. But in both cases his words fell on deaf ears. Bismarck had clearly decided to play the German national card in order to achieve a Prussian-dominated Germany. After making sure that Russia would not intervene and after gaining an alliance with Italy, he set about fostering conflict with the Austrians. He stirred up Hungarian nationalism against Austria—a policy that showed how radical means could be used in the service of his own conservative ends. On June 9, 1866, Prussian troops invaded Holstein, and a few days later Austria, supported by the smaller states of Saxony, Hesse-Kassel, and Hanover, went to war. Within six weeks Prussia had inflicted a major defeat on the Austrians at Königgrätz (Sadowa). Bismarck then counseled moderation so that Austria would not be humiliated. Against a king and generals who wanted to march to Vienna, he urged a quick cessation of hostilities, recognizing that other powers might intervene if the war continued. Europe was stunned: in a few weeks Prussia had transformed the distribution of power in central Europe. Austria, the major power in Germany for centuries, was now relegated to secondary status.

Bismarck now showed both ruthlessness and moderation. The Peace of Nikolsburg scarcely demanded anything from Austria. But Hanover, Hesse-Kassel, Nassau, and Frankfurt, all of which had fought against Prussia, were annexed, to the shock of conservatives. The king of Hanover was removed from power, as was the ruling house in Hesse. While conservatives were appalled at the German civil war between the two powers who had been opposed to revolution, the liberal middle class flocked to support Bismarck. Their goal of German unification seemed close at hand. Bismarck, moreover, now apologized for his high-handedness over the issue of the military budget and offered an olive branch of peace to the liberals. The party divided over Bismarck's offer. He had achieved one of his major goals—gaining a large part of the middle class to see the Prussian monarchy as their ally.

The North German Confederation was established in 1867 with Prussia as its matrix. Its constitution, on the

The North German Confederation

surface, appeared progressive. To begin with, it established universal manhood suffrage with a secret ballot. But this was a result of Bismarck's belief that the vast majority of Prussians, if enfranchised, would vote conservative. From this perspective, a restricted ballot aided the liberals. (Of course, in 1867 neither the socialists nor the Catholic Centre had established political parties.) Moreover, whereas in theory the lower house (Reichstag) seemed an important reservoir of power given its ability to reject any bill, in practice its powers were circumscribed in the areas of military and foreign policy. Ministers were chosen by and responsible to the king and not the legislature. Nevertheless, the constitution provided a basis for evolution in a democratic direction.

Although Bismarck voiced doubts whether unification would occur in his lifetime, he actually set about tying the southern states to the north almost immediately. An all-German customs parliament was proposed, joint military training was negotiated, and a plan was advanced which entailed that the southern states recognize William as German emperor. All these efforts failed because of popular opposition in the south. Bismarck then sought to propel history a bit faster by seeking conflict with France. If he could not bring the south into a united German nation by reason, he would rely on the passions aroused by war. Ever the master tactician, he worked behind the scenes to be certain that neither Russia nor Austria would intervene in such a war. Nor did he have to work hard to produce a conflict, because the French emperor, Napoleon III, was indignant at the sudden emergence of Prussia, especially since he did not receive the compensation he sought—the annexation of Luxembourg.

When in 1869 the Spanish throne was offered to the king's cousin, Prince Leopold of Hohenzollern-Sigmaringen, Napoleon III perceived this as an effort to encircle France. He twice sent his ambassador, Vincent Benedetti, to the Prussian king at Bad Ems, once to demand that acceptance of the offer be withdrawn (which it was on July 12) and a second time to demand that under no circumstances should a member of the Hohenzollern family accept the Spanish throne in the future. The king politely refused the second request. Bismarck received a telegram from Bad Ems (the Ems telegram) giving a detailed account of the interview between William I and the French ambassador, which he proceeded to edit and abridge for the press in such a way that the French appeared to seek a humiliation of the Prussian monarch, and the monarch's rejection of Napoleon's demands seemed insultingly brusque to the French. The French responded by declaring war on Prussia on July 19, 1870. When the French were decisively defeated at Sedan in September, it appeared as though Bismarck would be able to score a third rapid victory in seven years. But guerrilla warfare broke out, and Paris held out despite the capture of the emperor. Bismarck, however, stirred anti-French passions to such a fever pitch that in January 1871 the four southern states joined the North German Confederation to create the German Empire. The lesser German solution, with seven million German-speaking Austrians excluded, was the result of Bismarck's three wars. He was showered with honours and hailed as a national hero.

The German Empire

Imperial chancellor. It is important to note that the Germany Bismarck created was not the result of strong popular currents of nationalist sentiment but of cabinet diplomacy and war. Not all German-speaking areas of Europe were included but only as many as Prussia could unite while retaining hegemony. The new constitution was a revision of the Prussian constitution from 1867; it included the position of chancellor, designed with Bismarck specifically in mind. Bismarck also remained prime minister of Prussia until 1890, apart from a brief period in 1872–73.

The peace treaty with France was harsh. Alsace and Lorraine, two French provinces with sizable German-speaking populations, were annexed. Also, a five-billion-franc indemnity was exacted. While Austria and Denmark quickly forgot their defeats, France did not. Regardless of whether Bismarck annexed the provinces in response to German public opinion or for other reasons, French hostility was to haunt the German Empire until the provinces were returned to France in 1918.

Foreign policy. Until his resignation in 1890, Bismarck had a relatively free hand in the conduct of foreign policy. After three successful wars, he saw his task as promoting peace and gaining time so that the powerful German Empire would come to be accepted as natural. Bismarck's two areas of concern were the Balkans, where the disintegration of the Turkish empire could easily lead to conflict between the Habsburg monarchy and Russia, and France, where the desire to avenge the defeat at Sedan was strong. In each area a general European conflagration could flare up and involve Germany. In 1873 he embraced a pacific foreign policy when he negotiated the Dreikaiserbund (Three Emperors' League) with Russia and Austria-Hungary. But the alliance did not survive the Russo-Turkish War of 1877. When the Austrians and British threatened war over a Carthaginian peace imposed on Turkey by the Russian victors, Bismarck called for a peace congress in Berlin. The German chancellor succeeded in getting the Russians to moderate their gains, and peace was preserved.

But a European conflagration had barely been averted. Soon after the conference, Bismarck negotiated a defensive alliance with Austria-Hungary, which remained in effect through World War I. Although in the mid-1860s he had rejected such an alliance as harmful, he now considered it advantageous. Because he feared that the dissolution of the Habsburg monarchy would lead to Russian expansion into central Europe, he sought the alliance to gain leverage in Vienna. He steadfastly used it to prevent a war in the Balkans. In addition, he did not want seven million Austro-German Catholics seeking admission to the empire.

Dreikaiserbund

Having a solid ally, Bismarck demonstrated his virtuosity by negotiating a revived Dreikaiserbund in 1881. He now had influence in St. Petersburg as well as in Vienna to prevent a Balkan war. In 1882, Italy, fearing French hostility, joined the Dual Alliance, making it into the Triple Alliance. On the surface Bismarck had triumphed. France had no allies for a war of revenge, and, for the moment, a Balkan war seemed unlikely.

But the ephemeral nature of all these alliances soon became apparent. A crisis in Bulgaria inflamed Russo-Austrian relations, leading to a breakup of the revived league. Once again a war was avoided with Bismarck's intervention, but his efforts could not reconstitute the league. He then negotiated a separate secret treaty with Russia, while maintaining the 1879 accord with Austria-Hungary.

Between 1870 and 1890 Bismarck earned the respect of European leaders for his earnest efforts in behalf of peace. Apart from a few colonial acquisitions in the mid-1880s, Germany had acted as a satiate power. All of Bismarck's considerable tactical skills had been successful in creating a powerful German Empire in his first decade in power. For the next two decades these same skills maintained the peace.

Domestic policy. From the defeat of Austria in 1866 until 1878 Bismarck was allied primarily with the National Liberals. Together they created a civil and criminal code for the new empire and accomplished Germany's adoption of the gold standard and move toward free trade. Just as they had earlier written off Bismarck as an archconservative, liberals now viewed him as a comrade—a man who had rejected his conservative roots. Many conservative leaders agreed with this assessment. Bismarck had cashiered kings, gone to war against conservative regimes, and adopted policies that promoted rapid industrialization. Their fears were further enhanced when he joined liberals in a campaign against political Catholicism (Kulturkampf) in 1873.

Bismarck had not counted on the emergence of new parties such as the Catholic Centre or the Social Democrats, both of whom began participating in imperial and Prussian elections in the early 1870s. Along with the left liberal Progressive Party, he labeled them all enemies of the empire (*Reichsfeinde*). Each in its own way rejected his vision of a united Germany. The Progressives found the empire too conservative and its elite essentially feudal; the socialists questioned its capitalist character; and for the Centre the empire was Protestant and too centralized.

Bismarck's aim was clearly to destroy the Catholic Centre Party. He and the liberals feared the appeal of a clerical party to the one-third of Germans who professed Roman Catholicism. In Prussia the minister of public worship and education, Adalbert Falk, with Bismarck's blessing, introduced a series of bills establishing civil marriage, limiting the movement of the clergy, and dissolving religious orders. All church appointments were to be approved by the state. Clerical civil servants were purged from the Prussian administration. Hundreds of parishes and several bishoprics were left without incumbents.

The Kulturkampf failed to achieve its goals and, if anything, convinced the Catholic minority that their fear of persecution was real. Bismarck gradually relented in his campaign, especially after the death of the activist pope, Pius IX, in 1878. But he never relented in his hatred for the Centre leader, Ludwig Windthorst, a Hanoverian who had earlier experienced Bismarck's methods in the annexation of his kingdom. Bismarck's speeches continued to be barbed with anticlericalism until his fall in 1890.

Change in economic policy

In 1878–79 Bismarck initiated a significant change in economic policy, which coincided with his new alliance with the conservative parties at the expense of the liberals. Tariffs were introduced on iron as well as on major grains. The new policy was a result of the "great depression" that had swept Europe and the United States in the mid-1870s. Bismarck's shift had serious political implications: it signified his opposition to any further evolution in the direction of political democracy. The liberal ministers Falk and Rudolph von Delbrück resigned, and Robert von Puttkamer became minister of public worship and education in 1879 and minister of interior in 1881. The grain tariffs provided the Junker estate owners of Prussia, who constituted the main opposition to political reform, subventions that isolated them somewhat from the world market. From 1879 onward, the landed elite, major industrialists, the military, and higher civil servants formed an alliance to forestall the rise of social democracy.

Ever since the Commune of Paris of 1871, Bismarck had developed an uncompromising hatred for socialists and anarchists. His attacks on them were egregious. At one point he wrote, "They are this country's rats and should be exterminated." Another time he called them "a host of enemies bent on pillage and murder." He thus introduced a crude and unsavory discourse into everyday German politics that was to be long-lived. Although only two socialists sat in the Reichstag in 1871, their number and support grew with each election, until they had 35 seats in 1890. As early as 1876 Bismarck had sought legislation to outlaw the party but failed to get a majority. After two assassination attempts against William I he prorogued Parliament and ran a campaign in which the socialists (quite unjustly) were blamed for the failed efforts to kill the emperor. The conservative parties triumphed and the Social Democratic Party was banned in 1878. The ban was renewed until 1890.

The second part of Bismarck's strategy to destroy social democracy was the introduction of social legislation to woo the workers away from political radicalism. During the 1880s, accident and old-age insurance as well as a form of socialized medicine were introduced and implemented by the government. But Bismarck's two-pronged strategy to win the workers for the conservative regime did not succeed. Support for the Social Democrats increased with each election.

Resignation

The election of 1890 was a disaster for Bismarck. The Centre, the Social Democrats, and the Progressives, the parties that he had termed enemies of the empire, gained more than half of the seats in the new Reichstag. The new young emperor William II (b. 1859; emperor and king of Prussia from 1888 to 1918) did not want to begin his reign with a bloodbath or a coup d'état by the state. Seventy-five years old in 1890, Bismarck resigned with a sense of having failed. The antisocialist law was not revived, and the new government set out to win the workers to the regime. Bismarck retired to his estate an embittered man. That he was now a prince and extremely wealthy did not ease his retirement. For the next eight years (he died on July 30, 1898) he issued sharp critiques of his successors. Elected to the Reichstag, he chose not to take his seat. He wrote his memoirs, which became best-sellers. To some extent he orchestrated the Bismarck legend that was to dominate German historical writing for the next half century.

ASSESSMENT

Bismarck was a towering figure who put his stamp on his age, as Luther and Metternich had done earlier. When Bismarck became prime minister of Prussia in 1862, the kingdom was universally considered the weakest of the five European powers. Less than nine years later Prussia had been victorious in three wars, and a unified German Empire had emerged in the heart of Europe, arousing envy and fear among its rivals. When Bismarck left office in 1890, after 28 years as prime minister of Prussia and 19 as chancellor of the German Empire, the map of Europe had been changed beyond measure. The European centre, characterized by a weak conglomeration of small and medium-sized states for centuries, was now home to the foremost military and industrial power on the Continent.

Bismarck's legacy to the next generation, however, was a mixed one. In foreign affairs his skill had led to 20 years of peace in Europe, which had gained him a deserved reputation for moderation and a sense of limits. Bismarck's greatest achievement, the German Empire, only survived him by 20 years. Although he had united Germany in one sense, he had failed to create an internally unified people. In domestic affairs—as in foreign policy—he sought to freeze the status quo after 1871. His empire was designed to be conservative. Thus he opposed the Catholic Centre in the 1870s and the socialists in the 1880s because both constituted unforeseen threats to his authoritarian creation. He also introduced a vicious rhetoric into German politics that forestalled a sense of common destiny. While German industry developed rapidly during his decades in power, he would allow no evolution in the political system toward greater participation. In this sense, Bismarck was a last representative of the world of the ancien régime and cabinet diplomacy.

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Blood

Human blood consists of several types of specialized cells suspended in a liquid medium, plasma. The circulatory system provides the mechanism by which the blood transports many substances to and from the organs and tissues. The circulating blood continuously supplies oxygen, nutrient substances, and other materials necessary for the viability and activity of all the cells of the body and carries away cell products, including carbon dioxide and other waste materials. If blood flow ceases, death occurs within minutes because of the effects of an unfavourable environment on highly susceptible cells.

The constancy of the composition of the blood is made possible by the circulation, which conveys blood through the organs that regulate the concentrations of its components. In the lungs blood acquires oxygen and releases carbon dioxide transported from the tissues. The kidneys remove excess water and dissolved waste products. Nutrient substances derived from food reach the bloodstream after absorption by the intestinal tract. Endocrine glands release their secretions into the blood, which transports these hormones to the tissues in which they exert their effects. Many substances are recycled through the blood; for example, iron released during the destruction of old red cells is conveyed by the plasma to sites of new red cell production where it is reused. Each of the numerous components of the blood is kept within appropriate concentration limits by an effective regulatory mechanism. In many instances feedback control systems are operative; thus a declining level of blood sugar leads to accelerated release of sugar into the blood so that a potentially hazardous depletion of blood sugar does not occur.

Unicellular organisms, primitive multicellular animals, and the early embryos of higher forms of life lack a circulatory system, and exchange of substances between cell and environment is accomplished by simple diffusion. In larger and more complex animals transport of adequate amounts of oxygen and other substances requires some type of blood circulation. The diffusion process then occurs between the body cells and the fluid derived from the blood, which by its constant motion maintains the constancy of the internal environment. Some simple animals, including small worms and mollusks, have blood that lacks an oxygen-binding substance analogous to hemoglobin; others are provided with pigments capable of transporting relatively large amounts of oxygen. In many invertebrates the blood pigment is dissolved in the plasma. Hemocyanin, a copper-containing protein chemically unlike hemoglobin, is found in certain crabs and other lower animals. Hemocyanin is blue in colour when oxygenated and colourless when oxygen is removed. Some invertebrates have hemoglobin in solution in the plasma. In almost all vertebrates, including humans, hemoglobin is contained exclusively within the red cells (erythrocytes) of the blood. The red cells of the lower vertebrates (*e.g.,* birds) have a nucleus, whereas mammalian red cells lack a nucleus. Red cells vary markedly in size among mammals; those of the goat are much smaller than those of humans, but the goat compensates by having many more red cells per unit volume of blood. The concentration of hemoglobin inside the red cell varies little between species.

For additional information on blood in general and comparison of the blood and lymph of diverse organisms, see TISSUES AND FLUIDS: *The tissues and fluids of animals: Blood and lymph* and CIRCULATION AND CIRCULATORY SYSTEMS. For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 421, 422, and 423.

This article is divided into the following sections:

Characteristics of human blood

PROPERTIES

Blood is an opaque red fluid, freely flowing but denser and more viscous than water. The characteristic colour is imparted by hemoglobin, a unique iron-containing protein. Hemoglobin brightens in colour when saturated with oxygen (oxyhemoglobin) and darkens when oxygen is removed (deoxyhemoglobin). For this reason, the partially deoxygenated blood from a vein is darker than oxygenated blood from an artery. The red cells constitute about 45 percent of the volume of the blood, and the remaining cells (white cells and platelets) less than 1 percent. The fluid portion, plasma, is a clear, slightly sticky, yellowish liquid. After a fatty meal plasma transiently appears turbid. Within the body the blood is permanently fluid, and turbulent flow assures that cells and plasma are fairly homogeneously mixed.

Characteristics of shed blood

When blood is shed, physicochemical changes are initiated that cause the blood to coagulate (see below *Bleeding and blood clotting*). The blood clot consists of microscopic strands of a complex protein, called fibrin, forming a gel in which the erythrocytes and other cells are entrapped. When the clot shrinks, or retracts, it squeezes out an incoagulable yellowish fluid, which is called the serum. An anticoagulant can be added to the shed blood to prevent clot formation, thereby maintaining the blood in a fluid state. When blood treated in this way is undisturbed, the cells gradually settle because they are denser than the plasma; the red cells go to the bottom, the white cells and platelets form a thin white layer (buffy coat) overlying the red cells, and the plasma appears in the upper portion of the container. Rapid segregation of cells and plasma may be accomplished by the use of a centrifuge, a machine in which rapid rotation accelerates sedimentation by increasing gravitational forces.

The total amount of blood varies with age, sex, weight, body build, and other factors, but a rough average figure for adults is about 60 millilitres per kilogram of body weight. An average young male has a plasma volume of about 35 millilitres, and a red cell volume of about 30 millilitres, per kilogram of body weight. There is little variation in the blood volume of a healthy person over long periods, although each component of the blood is in a continuous state of flux. In particular, water rapidly moves in and out of the bloodstream, achieving a balance with the extravascular fluids (those outside the blood vessels) within minutes. The normal volume of blood provides such an adequate reserve that appreciable blood loss is well tolerated. Withdrawal of 500 millilitres (about a pint) of blood from normal blood donors is a harmless procedure. Blood volume is rapidly replaced after blood loss; within hours plasma volume is restored by movement of extravascular fluid into the circulation. Replacement of red cells is completed within several weeks. The vast area of capillary membrane, through which water passes freely, would permit instantaneous loss of the plasma from the circulation were it not for the plasma proteins, in particular, the plasma albumin. Capillary membranes are impermeable to albumin, the smallest in weight and highest in concentration of the plasma proteins. The osmotic effect of the plasma albumin retains fluid within the circulation, opposing the hydrostatic forces that tend to drive the fluid outward into the tissues.

FUNCTIONS

Maintaining homeostasis

Broadly conceived, the function of the blood is to maintain the constancy of the internal environment (homeostasis). The circulating blood makes possible human adaptability to changing conditions of life—the endurance of wide variations of climate and atmospheric pressure; the capacity to alter the amount of physical activity; the tolerance of changing diet and fluid intake; the resistance to physical injury, chemical poisons, and infectious agents. The blood has an exceedingly complex structure, and many components participate in its functional activities. Some of the regulatory mechanisms with which the blood is involved include sensors that detect alterations in temperature, in pH, in oxygen tension, and in concentrations of the constituents of the blood. Effects of these stimuli are in some instances mediated via the nervous system or by the release of hormones (chemical mediators). Some of the major functions of the blood are outlined in the paragraphs that follow.

pH. The pH of blood is kept relatively constant at the slightly alkaline level of about 7.4 (a pH of less than 7 indicates acidity, of more than 7 alkalinity). Venous blood is maintained at a somewhat less alkaline level (7.35) because of the higher carbon dioxide content. A system of efficient buffers in the blood and the selective excretory functions of the lungs and kidneys keep the pH within these narrow limits. Physiological mechanisms stabilize a normal pH in the blood both by regulating the rate and depth of respiration in order to maintain a normal tension of carbon dioxide in the blood and by excreting acid or alkaline urine from the kidneys.

Maintenance of pH

Respiration. In terms of immediate urgency, the respiratory function of the blood is vital. A continuous supply of oxygen is required by living cells, in particular those of the brain since deprivation is followed in minutes by unconsciousness and death. A normal male at rest uses about 250 millilitres of oxygen per minute, a requirement increased manyfold during vigorous exertion. All of this oxygen is transported by the blood, most of it bound to the hemoglobin of the red cells. The minute blood vessels of the lungs bring the blood into close apposition with the pulmonary air spaces (alveoli), where the pressure of oxygen is relatively high. Oxygen diffuses through the plasma and into the red cell, combining with hemoglobin, which is about 95 percent saturated with oxygen on leaving the lungs. One gram of hemoglobin can bind 1.35 millilitres of oxygen, and about 50 times as much oxygen is combined with hemoglobin as is dissolved in the plasma. In tissues where the oxygen tension is relatively low, hemoglobin releases the bound oxygen.

The two main regulators of oxygen uptake and delivery are the pH of tissues and the content of 2,3-diphosphoglycerate (2,3-DPG) in red cells. The effect of pH on the ability of hemoglobin to bind oxygen is called the Bohr effect (Figure 1): when pH is low, hemoglobin binds oxygen less strongly, and when pH is high (as in the lungs), hemoglobin binds more tightly to oxygen. The Bohr effect is due to changes in the shape of the hemoglobin molecule as the pH of its environment changes. The oxygen affinity of hemoglobin is also regulated by 2,3-DPG, a simple molecule produced by the red cell when it metabolizes glucose. The effect of 2,3-DPG is to reduce the oxygen affinity of hemoglobin. When the availability of oxygen to tissues is reduced, the red cell responds by synthesizing more 2,3-DPG, a process that occurs over a period of hours to days. By contrast, tissue pH mediates minute-to-minute changes in oxygen handling.

Elimination of carbon dioxide

Carbon dioxide, a waste product of cellular metabolism, is found in relatively high concentration in the tissues. It diffuses into the blood and is carried to the lungs to be eliminated with the expired air. Carbon dioxide is much more soluble than oxygen and readily diffuses into red cells. It reacts with water to form carbonic acid, a weak

From A.L. Lehninger, *Principles of Biochemistry* (1982), Worth Publishers, Inc., New York City

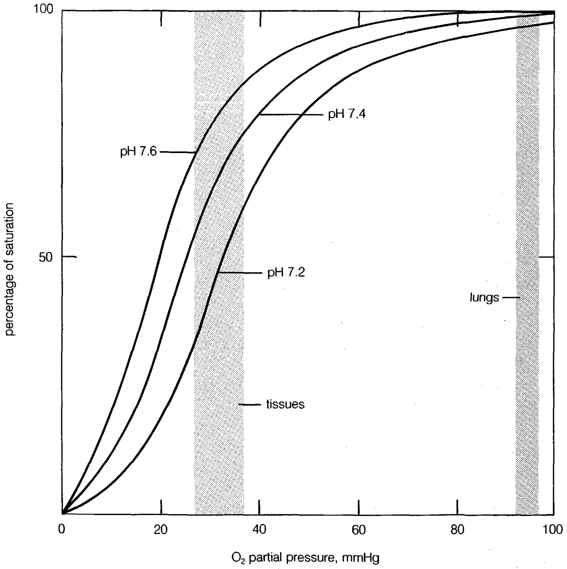

Figure 1: *The Bohr effect.*
The oxygen affinity of hemoglobin weakens with decreasing pH. The body tissue has a lower pH, and the oxygen is released more readily. In the lungs, the pH is higher and oxygen is more easily bound to the hemoglobin molecule.

acid that at the alkaline pH of the blood appears principally as bicarbonate.

The tension of carbon dioxide in the arterial blood is regulated with extraordinary precision through a sensing mechanism in the brain that controls the respiratory movements. Carbon dioxide is an acidic substance, and an increase in its concentration tends to lower the pH of the blood (*i.e.,* becoming more acidic). This may be averted by the stimulus that causes increased depth and rate of breathing, a response that accelerates the loss of carbon dioxide. It is the tension of carbon dioxide, and not of oxygen, in the arterial blood that normally controls breathing. Inability to hold one's breath for more than a minute or so is the result of the rising tension of carbon dioxide, which produces the irresistible stimulus to breathe. Respiratory movements that ventilate the lungs sufficiently to maintain a normal tension of carbon dioxide are, under normal conditions, adequate to keep the blood fully oxygenated. Control of respiration is effective, therefore, in regulating the uptake of oxygen and disposal of carbon dioxide and in maintaining the constancy of blood pH.

Nutrition. Each substance required for the nutrition of every cell in the body is transported by the blood: the precursors of carbohydrates, proteins, and fats; minerals and salts; vitamins and other accessory food factors. These substances must all pass through the plasma on the way to the tissues in which they are used. The materials may enter the bloodstream from the gastrointestinal tract, or they may be released from stores within the body or become available from the breakdown of tissue.

Regulators of nutrients

The concentrations of many plasma constituents, including glucose and calcium, are carefully regulated, and deviations from the normal may have adverse effects. One of the regulators of the blood sugar is insulin, a hormone released into the blood from glandular cells in the pancreas. Ingestion of a carbohydrate meal is followed by increased production of insulin, which tends to keep the blood sugar from rising excessively as the carbohydrates are broken down into their constituent sugar molecules. But an excess of insulin may severely reduce the blood sugar, causing a reaction that, if sufficiently severe, may include coma and even death. Glucose is transported in simple solution, but some substances require specific binding proteins (proteins with which the substances form temporary unions) to convey them through the plasma. Iron and copper, essential minerals, have special and necessary transport proteins. Nutrient substances may be taken up selectively by the tissues that require them. Growing bones use large amounts of calcium, and bone marrow removes iron from plasma for hemoglobin synthesis.

Excretion. The blood carries the waste products of cellular metabolism to the excretory organs. The removal of carbon dioxide via the lungs has been described above. Water produced by the oxidation of foods or available from other sources in excess of needs is excreted by the kidneys as the solvent of the urine. Water derived from the blood also is lost from the body by evaporation from the skin and lungs and in small amounts from the gastrointestinal tract. The water content of the blood and of the body as a whole remains within a narrow range because of effective regulatory mechanisms, hormonal and other, that determine the urinary volume. The concentrations of physiologically important ions of the plasma, notably sodium, potassium, and chloride, are precisely controlled by their retention or selective removal as blood flows through the kidneys. Of special significance is the renal (kidney) control of acidity of the urine, a major factor in the maintenance of the normal pH of the blood. Urea, creatinine, and uric acid are nitrogen-containing products of metabolism that are transported by the blood and rapidly eliminated by the kidneys. The kidneys clear the blood of many other substances, including numerous drugs and chemicals that are taken into the body. In performing their excretory function, the kidneys have a major responsibility for maintaining the constancy of the composition of the blood. (See also EXCRETION AND EXCRETORY SYSTEMS.) The liver is in part an excretory organ. Bilirubin (bile pigment) produced by the destruction of hemoglobin is conveyed by the plasma to the liver and is excreted through the biliary ducts into the intestinal tract. Other substances, including certain drugs, also are removed from the plasma by the liver.

Immune and inflammatory responses

Defense mechanisms. Cells of the blood and constituents of the plasma interact in complex ways to confer immunity to infectious agents, to resist or destroy invading organisms, to produce the inflammatory response, and to destroy and remove foreign materials and dead cells. The leukocytes (white blood cells, discussed later in this section) have a primary role in these reactions: granulocytes and monocytes phagocytize (ingest) bacteria and other organisms, migrate to sites of infection or inflammation and to areas containing dead tissue, and participate in the enzymatic breakdown and removal of cellular debris; lymphocytes are concerned with the development of immunity. Acquired resistance to specific microorganisms is in part attributable to antibodies, proteins that are formed in response to the entry into the body of a foreign substance (antigen). Antibodies that have been induced by microorganisms not only participate in eliminating the microbes but also prevent reinfection by the same organism. Cells and antibodies may cooperate in the destruction of invading bacteria; the antibody may attach to the organism, thereby rendering it susceptible to phagocytosis. Involved in some of these reactions is complement, a group of protein components of plasma that participates in certain immunologic reactions. When certain classes of antibodies bind to microorganisms and other cells, they trigger the attachment of components of the complement system to the outer membrane of the target cell. As they assemble on the cell membrane, the complement components acquire enzymatic properties. The activated complement system is thus able to injure the cell by digesting (lysing) portions of the cell's protective membrane.

Hemostasis. The blood is contained under pressure in a vascular system that includes vast areas of thin and delicate capillary membranes. Even the bumps and knocks of everyday life are sufficient to disrupt some of these fragile vessels, and serious injury can be much more damaging. Loss of blood would be a constant threat to survival if it were not for protective mechanisms to prevent and to control bleeding. The platelets contribute to the resistance of capillaries, possibly because they actually fill chinks in vessel walls. In the absence of platelets capillaries become more fragile, permitting spontaneous loss of blood and increasing the tendency to form bruises after minor injury. Platelets immediately aggregate at the site of injury of a blood vessel, tending to seal the aperture. A blood clot, forming in the vessel around the clump of adherent platelets, further occludes the bleeding point. The coagulation mechanism involves a series of chemical reactions in which specific proteins and other constituents of the blood, including the platelets, play a part. Plasma also is provided with a mechanism for dissolving clots after they have been formed. Plasmin is a proteolytic enzyme—a substance that causes breakdown of proteins—derived from an inert plasma precursor known as plasminogen. When clots are formed within blood vessels, activation of plasminogen to plasmin may lead to their removal. (For additional information about the mechanics and significance of hemostasis, see below *Bleeding and blood clotting.*)

Mechanisms to prevent blood loss

Temperature regulation. Heat is produced in large amounts by physiological oxidative reactions, and the blood is essential for its distributing and disposing of this heat. The circulation assures relative uniformity of temperature throughout the body and also carries the warm blood to the surface, where heat is lost to the external environment. A heat-regulating centre in the hypothalamus of the brain functions much like a thermostat. It is sensitive to changes in temperature of the blood flowing through it and, in response to the changes, gives off nerve impulses that control the calibre of the blood vessels in the skin and thus determine blood flow and skin temperature. A rise in skin temperature increases heat loss from the body surface. Heat is continuously lost by evaporation of water from the lungs and skin, but this loss can be greatly increased when more water is made available from the sweat glands. The activity of the sweat glands is controlled by the nervous system under direction of the temperature-regulating cen-

The hypothalamus

tre. Constancy of body temperature is achieved by control of the rate of heat loss by these mechanisms.

PLASMA

The liquid portion of the blood, the plasma, is a complex solution containing more than 90 percent water. The water of the plasma is freely exchangeable with that of body cells and other extracellular fluids and is available to maintain the normal state of hydration of all tissues. Water, the single largest constituent of the body, is essential to the existence of every living cell. The major solute of plasma is a heterogeneous group of proteins constituting about 7 percent of the plasma by weight. The principle difference between the plasma and the extracellular fluid of the tissues is the high protein content of the plasma. Plasma protein exerts an osmotic effect by which water tends to move from other extracellular fluid to the plasma. Fatty substances (lipids) are present in plasma in suspension and in solution. Other plasma constituents include salts, glucose, amino acids, vitamins, hormones, and waste products of metabolism.

Proteins. Proteins are large molecules formed of chains of amino acids, organic acids that contain both an acidic and a nitrogenous basic (amino) group. Chains (polypeptides) are formed by linkage of the acid group of one amino acid to the amino group of the next (peptide bond). The characteristics of a protein are determined by the number and types of amino acids and the sequence in which they are arranged.

Albumin

When dietary protein is digested in the gastrointestinal tract, individual amino acids are released from the chains and are absorbed. The amino acids are transported through the plasma to all parts of the body, where they are taken up by cells and are assembled in specific ways to form proteins of many types. These plasma proteins are released into the blood from the cells in which they were synthesized. Much of the protein of plasma is produced in the liver. The major plasma protein is albumin, a relatively small molecule, the principal function of which is to retain water in the bloodstream by its osmotic effect. The amount of albumin in the blood is a determinant of the total volume of plasma. Depletion of albumin permits fluid to leave the circulation and to accumulate and cause swelling of soft tissues (edema). Albumin binds certain other substances that are transported in plasma and thus serves as a nonspecific carrier protein. Bilirubin, for example, is bound to albumin during its passage through the blood.

Albumin has physical properties that permit its separation from other plasma proteins, which as a group are called globulins. In fact, the globulins are a heterogeneous array of proteins of widely varying structure and function, only a few of which will be mentioned here. The immunoglobulins, or antibodies, are produced in response to a specific antigen. For example, administration of poliomyelitis vaccine is followed by the appearance in the plasma of antibodies that react with poliovirus and effectively prevent that infection. Antibodies may be induced by many foreign substances in addition to microorganisms; immunoglobulins are involved in some hypersensitivity and allergic reactions. Other plasma proteins are concerned with the coagulation of the blood (see below *Bleeding and blood clotting*).

Transport proteins

Many proteins are involved in highly specific ways with the transport function of the blood. Blood lipids are incorporated into protein molecules as lipoproteins, substances important in lipid transport. Iron and copper are transported in plasma by unique metal-binding proteins (transferrin and ceruloplasmin, respectively). Vitamin B_{12}, an essential nutrient, is bound to a specific carrier protein. Although hemoglobin is not normally released into the plasma, a hemoglobin-binding protein (haptoglobin) is available to transport hemoglobin should hemolysis (breakdown) of red cells occur.

Lipids. The concentration of lipids in plasma varies, particularly in relation to meals, but ordinarily does not exceed one gram per 100 millilitres. The largest fraction consists of phospholipids, complex molecules containing phosphoric acid and a nitrogen base in addition to fatty acids and glycerol. Triglycerides, or simple fats, are molecules composed only of fatty acids and glycerol. Free fatty acids, lower in concentration than triglycerides, are responsible for a much larger transport of fat. Other lipids include cholesterol, a major fraction of the total plasma lipids. These substances exist in plasma combined with proteins of several types as lipoproteins. The largest lipid particles in the blood are known as chylomicrons and consist largely of triglycerides; after absorption from the intestine they pass through lymphatic channels and enter the bloodstream through the thoracic lymph duct. The other plasma lipids are derived from food or enter the plasma from tissue sites.

Other plasma components. Some plasma constituents occur in plasma in low concentration but have a high turnover rate and great physiological importance. Among these is glucose, the blood sugar. Glucose is absorbed from the gastrointestinal tract or may be released into the circulation from the liver. It provides a source of energy for tissue cells and is the only source for some, including the red cells. Glucose is conserved and used and not excreted. Amino acids also are so rapidly transported that the plasma level remains low, although they are required for all protein synthesis throughout the body. Urea, an end product of protein metabolism, is rapidly excreted by the kidneys. Other nitrogenous waste products—uric acid and creatinine—are similarly removed.

Sodium, potassium, and other metals

Several inorganic materials are essential constituents of plasma, and each has special functional attributes. The predominant cation (positively charged ion) of the plasma is sodium, an ion that occurs within cells at a much lower concentration. Because of the effect of sodium on osmotic pressure and fluid movements, the amount of sodium in the body is an influential determinant of the total volume of extracellular fluid. The amount of sodium in plasma is controlled by the kidneys under the influence of a hormone (aldosterone) of the adrenal gland. If dietary sodium exceeds requirements, the excess is excreted by the kidneys. Potassium, the principal intracellular cation, occurs in plasma at a much lower concentration than sodium. The renal excretion of potassium is influenced by aldosterone, which causes retention of sodium and loss of potassium. Calcium in plasma is in part bound to protein and in part ionized. Its concentration is under the control of two hormones: parathyroid hormone, which causes the level to rise, and calcitonin, which causes it to fall. Magnesium, like potassium, is a predominantly intracellular cation and occurs in plasma in low concentration. Variations in the concentrations of these cations may have profound effects on the nervous system, the muscles, and the heart, effects normally prevented by precise regulatory mechanisms. Iron, copper, and zinc are required in trace amounts for synthesis of essential enzymes; much more iron is needed in addition for production of hemoglobin and myoglobin, the oxygen-binding pigment of muscles. These metals occur in plasma in low concentrations. The principal anion (negatively charged ion) of plasma is chloride; sodium chloride is its major salt. Bicarbonate participates in the transport of carbon dioxide and in the regulation of pH. Phosphate also has a buffering effect on the pH of the blood and is vital for chemical reactions of cells and for the metabolism of calcium. Iodide is transported through plasma in trace amounts; it is avidly taken up by the thyroid gland, which incorporates it into thyroid hormone.

The hormones of all the endocrine glands are secreted into the plasma and transported to their target organs, the organs in which they exert their effect. The plasma levels of these agents often reflect the functional activity of the glands that secrete them; in some instances measurements are possible though concentrations are extremely low. Among the many other constituents of plasma are numerous enzymes. Some of these appear simply to have escaped from tissue cells and have no functional significance in the blood.

BLOOD CELLS

There are four major types of blood cells: red blood cells, platelets, lymphocytes, and phagocytic cells. Collectively, Functions of the blood cells

the lymphocytes and phagocytic cells constitute the white blood cells. Each type of blood cell has a specialized function: red cells take up oxygen from the lungs and deliver it to the tissues; platelets participate in forming blood clots; lymphocytes are involved with immunity; and phagocytic cells occur in two varieties—granulocytes and monocytes—and ingest and break down microorganisms and foreign particles. The circulating blood functions as a conduit, bringing the various kinds of cells to the regions of the body in which they are needed: red cells to tissues requiring oxygen, platelets to seal over points of injury, lymphocytes to areas of infection, and phagocytic cells to sites of microbial invasion and inflammation. Each type of blood cell is described in detail below.

The process of blood cell formation (hematopoiesis) takes place in hematopoietic tissue. In the developing embryo, blood cell formation occurs in the liver, but, as the fetus develops, hematopoiesis shifts to the bone marrow, a dark red, gelatinous tissue in the central cavities of the bones. In young children, hematopoietic bone marrow fills most of the skeleton, whereas in adults the marrow is located mainly in the central bones (ribs, sternum, vertebrae, and pelvic bones). Bone marrow is a rich mixture of developing and mature blood cells, as well as fat cells and other cells that provide nutrition and an architectural framework upon which the blood-forming elements arrange themselves. The weight of the marrow of a normal adult is 1,600 to 3,700 grams and contains over 1,000,000,000,000 hematopoietic cells (18×10^9 cells per kilogram). Nourishment of this large mass of cells comes from the blood itself. Arteries pierce the outer walls of the bones, enter the marrow, and divide into fine branches, which ultimately coalesce into large venous sacs (sinusoids) through which blood flows sluggishly. In the surrounding hematopoietic tissue, newly formed blood cells enter the general circulation by penetrating the walls of the sinusoids.

Colony-stimulating factors

All blood cells arise from primordial cells called multipotent hematopoietic stem cells. By dividing and differentiating, these precursor cells give rise to the four major blood cell lineages: red cells, phagocytic cells, megakaryocytes, and lymphocytes. The cells of the marrow are under complex controls that regulate their formation and adjust their production to the changing demands of the body. When marrow stem cells are cultured outside the body they form tiny clusters of cells (colonies), which correspond to red cells, phagocytic cells, and megakaryocytes. The formation of these individual colonies depends on hormonal sugar-containing proteins (glycoproteins), referred to collectively as colony-stimulating factors (CSFs). These factors are produced throughout the body. Even in minute amounts, CSFs can stimulate the division and differentiation of precursor cells into mature blood cells and thus exert powerful regulatory influences over the production of blood cells. A master colony-stimulating factor (multi-CSF), also called interleukin-3, stimulates the most ancestral hematopoietic stem cell. Further differentiation of this stem cell into specialized descendants requires particular kinds of colony-stimulating factors; for example, the CSF erythropoietin is needed for the maturation of red cells, and granulocyte colony-stimulating factor controls the production of granulocytes. These glycoproteins, as well as other colony-stimulating factors, serve as signals from the tissues to the marrow. For instance, a decrease in the oxygen content of the blood stimulates the kidney to increase its production of erythropoietin, thus ultimately raising the number of oxygen-carrying red cells in the blood. Certain bacterial components accelerate the formation of granulocyte colony-stimulating factor, thereby leading to an increased production of phagocytic granulocytes by the bone marrow during infection.

Hematocrit

Red blood cells (erythrocytes). The red blood cells are highly specialized, well adapted for their primary function of transporting oxygen from the lungs to all of the body tissues. Red cells are approximately 7.8 micrometres in diameter and have the form of biconcave disks, a shape that provides a large surface-to-volume ratio. When blood is centrifuged to cause the cells to settle, the volume of packed red cells (hematocrit value) ranges between 42 and 54 percent of total volume in men and between 37 and 47 percent in women; values are somewhat lower in children. Normal red blood cells are fairly uniform in volume, so that the hematocrit value is determined largely by the number of red cells per unit of blood. The normal red cell count ranges between 4,000,000 and 6,000,000 per cubic millimetre. Hemoglobin constitutes about one-third of the weight of each red cell. The amount of hemoglobin in blood is related to the hematocrit value and to the red cell count, and in normal adults ranges between 14 and 18 grams per 100 millilitres. When fresh blood is examined with the microscope, red cells appear to be yellow-green disks with pale centres containing no visible internal structures.

The red cell is enclosed in a thin membrane that is composed of chemically complex lipids, proteins, and carbohydrates in a highly organized structure. Extraordinary distortion of the red cell occurs in its passage through minute blood vessels, many of which have a diameter less than that of the red cell. When the deforming stress is removed, the cell springs back to its original shape. The red cell readily tolerates bending and folding, but, if appreciable stretching of the membrane occurs, the cell is damaged or destroyed. The membrane is freely permeable to water, oxygen, carbon dioxide, glucose, urea, and certain other substances, but it is impermeable to hemoglobin. Within the cell the major cation is potassium; in contrast, in plasma and extracellular fluids the major cation is predominantly sodium. A pumping mechanism, driven by enzymes within the red cell, maintains its sodium and potassium concentrations. Red cells are subject to osmotic effects. When they are suspended in very dilute (hypotonic) solutions of sodium chloride, red cells take in water, which causes them to increase in volume and to become more spheroid; in concentrated salt solutions they lose water and shrink. In distilled water red cells continue to swell until they become spherical, whereupon they disrupt, releasing the dissolved hemoglobin into the surrounding fluid (hemolysis).

Breakdown of red cells

Hemolysis. When red cell membranes are damaged, hemoglobin and other dissolved contents may escape from the cells, leaving the membranous structures as "ghosts." This process, called hemolysis, is produced not only by the osmotic effects of water but by numerous other mechanisms. These include physical damage to red cells, as when blood is heated, is forced under great pressure through a small needle, or is subjected to freezing and thawing; chemical damage to red cells by agents such as bile salts, detergents, and certain snake venoms; and damage caused by immunologic reactions that may occur when antibodies attach to red cells in the presence of complement. When such destruction proceeds at a greater than normal rate, hemolytic anemia results (see below *Blood diseases*).

Blood groups. The membrane of the red cell has on its surface a group of molecules that confer blood group specificity (*i.e.,* that differentiate blood cells into groups). Most blood group substances are composed of carbohydrate linked to protein, and it is usually the chemical structure of the carbohydrate portion that determines the specific blood type. Blood group substances are antigens capable of inducing the production of antibodies when injected into persons or animals lacking the antigen. Detection and recognition of the blood group antigens are accomplished by the use of serum containing these antibodies. The large number of different red cell antigens makes it extremely unlikely that persons other than identical twins will have the same array of blood group substances. (For a full treatment of the subject, see below *Blood groups.*)

Interaction of hemoglobin and oxygen

Hemoglobin. About 95 percent of the dry weight of the red cell consists of hemoglobin, the substance necessary for oxygen transport. Hemoglobin is a protein; a molecule contains four polypeptide chains (a tetramer), each chain consisting of more than 140 amino acids (Figure 2). To each chain is attached a chemical structure known as a heme group. Heme is composed of a ringlike organic compound known as a porphyrin to which an iron atom is attached. It is the iron atom that reversibly binds oxygen as the blood travels between the lungs and the tissues. There are four iron atoms in each molecule of hemoglobin, which, accordingly, can bind four atoms of oxygen. The complex porphyrin and protein structure may be considered to provide just the proper environment for the iron atom so that it binds and releases oxygen appropriately under physiological conditions. The affinity of hemoglobin for oxygen is so great that at the oxygen pressure in the lungs about 95 percent of the hemoglobin is saturated with oxygen. As the oxygen tension falls, as it does in the tissues, oxygen dissociates from hemoglobin and is available to move by diffusion through the red cell membrane and the plasma to sites where it is used. The proportion of hemoglobin saturated with oxygen is not directly proportional to the oxygen pressure. As the oxygen pressure declines, hemoglobin gives up its oxygen with disproportionate rapidity, so that the major fraction of the oxygen can be released with a relatively small drop in oxygen tension. The affinity of hemoglobin for oxygen is primarily determined by the structure of hemoglobin, but it is also influenced by other conditions within the red cell, in particular the pH and certain organic phosphate compounds produced during the chemical breakdown of glucose, especially 2,3-diphosphoglycerate (see above *Functions*).

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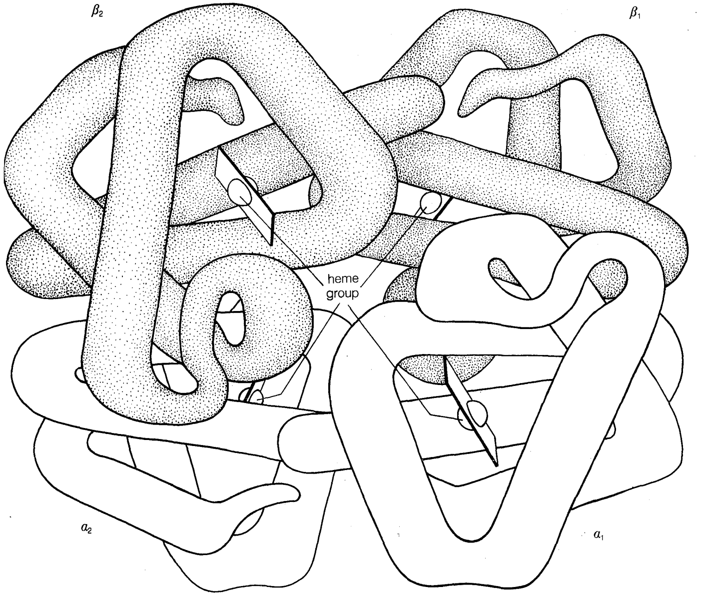

Figure 2: *The hemoglobin tetramer.*
Two $\alpha\beta$ dimers combine to form the complete hemoglobin molecule. Each heme group contains a central iron atom, which is available to bind a molecule of oxygen. The $\alpha_1\beta_2$ region is the area where the α_1 subunit interacts with the β_2 subunit.

Hemoglobin has a much higher affinity for carbon monoxide than for oxygen. Carbon monoxide produces its lethal effects by binding to hemoglobin and preventing oxygen transport. The oxygen-carrying function of hemoglobin can be disturbed in other ways. The iron of hemoglobin is normally in the reduced or ferrous state, both in oxyhemoglobin and deoxyhemoglobin. If the iron itself becomes oxidized to the ferric state, hemoglobin is changed to methemoglobin, a brown pigment incapable of transporting oxygen. The red cells contain enzymes capable of maintaining the iron in its normal state, but under abnormal conditions large amounts of methemoglobin may appear in the blood.

Sickle-cell anemia

Discovery of the cause of sickle-cell anemia has led to major advances in understanding of genetics, molecular biology, and the mechanisms of disease. Sickle-cell anemia is a serious and often fatal disease characterized by an inherited abnormality of the hemoglobin. Persons who have sickle-cell anemia are predominantly blacks. The disease is caused by the mutation of a single gene that determines the structure of the hemoglobin molecule. Sickle hemoglobin differs from normal hemoglobin in that a single amino acid (glutamic acid) in one pair of the polypeptide chains has been replaced by another (valine). This single intramolecular change so alters the properties of the hemoglobin molecule that anemia and other effects are produced. The entire structure of the hemoglobin molecule is known, and many other genetically determined abnormalities have been identified. Some of these also produce diseases of several types. Study of the effects of altered structure of hemoglobin on its properties has greatly broadened knowledge of the structure-function relationships of the hemoglobin molecule. (For more information about sickle-cell anemia, see below *Blood diseases.*)

Red cell metabolism. Survival of the red cell in the circulation depends upon the continuous utilization of glucose for the production of energy. Two chemical pathways are employed, and both are essential for the normal life of the red cell. An extraordinary number of enzyme systems participate in these reactions and direct the energy evolved into appropriate uses. Red cells contain neither a nucleus nor RNA (ribonucleic acid, necessary for protein synthesis), so that cell division and production of new protein are impossible. Energy is not necessary for oxygen and carbon dioxide transport, which depends principally on the properties of hemoglobin. Energy, however, is needed for another operation. There is a tendency for the extracellular cation, sodium, to leak into the red cell and for potassium to leak out; energy is required to operate a pumping mechanism in the red cell membrane to maintain the normal gradients (differences in concentrations) of these ions. Energy is also required to convert methemoglobin to oxyhemoglobin and to prevent the oxidation of other constituents of the red cell.

Erythropoiesis (production of red cells). Red cells are produced continuously in the marrow of certain bones. As stated above, in adults the principal sites are the marrow spaces of the vertebrae, ribs, breastbone, and pelvis. Within the bone marrow the red cell is derived from a primitive precursor, or erythroblast, a nucleated cell in which there is no hemoglobin. Proliferation occurs as a result of several successive cell divisions. During maturation hemoglobin appears in the cell, and the nucleus becomes progressively smaller. After a few days the cell loses its nucleus and is then introduced into the bloodstream in the vascular channels of the marrow. Almost 1 percent of the red cells are generated each day, and the balance between red cell production and the removal of aging red cells from the circulation is precisely maintained. If blood is lost from the circulation, the erythropoietic activity of marrow increases until the normal number of circulating cells has been restored.

Nutrients needed for red cell production

In a normal adult the red cells of about half a litre (almost one pint) of blood are produced by the bone marrow every week. A number of nutrient substances are required for this process. Some nutrients are the building blocks of which the red cells are composed. For example, amino acids are needed in abundance for the construction of the proteins of the red cell, in particular of hemoglobin. Iron also is a necessary component of hemoglobin. Approximately one-quarter of a gram of iron is needed for the production of a pint of blood. Other substances, required in trace amounts, are needed to catalyze the chemical reactions by which red cells are produced. Important among these are several vitamins, riboflavin, vitamin B_{12}, and folic acid, necessary for the maturation of the developing red cell; and pyridoxine (vitamin B_6), required for the synthesis of hemoglobin. The secretions of several endocrine glands influence red cell production. If there is an inadequate supply of thyroid hormone, erythropoiesis is retarded and anemia appears. The male sex hormone, testosterone, stimulates red cell production; for this reason, red cell counts of men are higher than those of women.

The capacity of the bone marrow to produce red cells is enormous. When stimulated to peak activity and when provided adequately with nutrient substances, the marrow can compensate for the loss of several pints of blood per week. Hemorrhage or accelerated destruction of red cells leads to enhanced marrow activity. The marrow can increase its production of red cells up to eight times the usual rate. After that, if blood loss continues, anemia develops. The rate of erythropoiesis is sensitive to the oxygen tension of the arterial blood. When oxygen tension falls, more red cells are produced and the red cell count rises.

Effects of altitude

For this reason, persons who live at high altitude have higher red cell counts than those who live at sea level. There is a small but significant difference between average red cell counts of persons living in New York City, at sea level pressure, and persons living in Denver, Colo., one mile above sea level, where the atmospheric pressure is lower. Natives of the Andes, living nearly three miles above sea level, have extremely high red cell counts.

The rate of production of erythrocytes is controlled by a hormone (erythropoietin) that is produced largely in the kidneys. When the number of circulating red cells decreases or when the oxygen transported by the blood diminishes, an unidentified sensor detects the change and the production of erythropoietin is increased. This substance is then transported through the plasma to the bone marrow, where it accelerates the production of red cells. The erythropoietin mechanism operates like a thermostat, increasing or decreasing the rate of red cell production in accordance with need. When a person who has lived at high altitude moves to a sea level environment, production of erythropoietin is suppressed, the rate of red cell production declines, and the red cell count falls until the normal sea level value is achieved. With the loss of one pint of blood, the erythropoietin mechanism is activated, red cell production is enhanced, and within a few weeks the number of circulating red cells has been restored to the normal value. The precision of control is extraordinary, so that the number of new red cells produced accurately compensates for the number of cells lost or destroyed. Erythropoietin has been produced in vitro (outside the body) by the technique of genetic engineering (recombinant DNA). The purified, recombinant hormone has promise for persons with chronic renal failure, who develop anemia because of a lack of erythropoietin.

Breakdown and removal

Destruction of red cells. Red cells have an average life span of 120 days. Although they use glucose to produce energy necessary for their survival, they cannot synthesize protein; therefore reparative processes are not possible. As red cells age, wear and tear leads to loss of some of the protein, and the activity of some of the essential enzymes decreases. Chemical reactions necessary for the survival of the cell are consequently impaired. As a result, water passes into the aging red cell, transforming its usual discoid shape into a sphere. These spherocytes are inelastic, and, as they sluggishly move through the circulation, they are engulfed by phagocytes (Figure 3). Phagocytic cells form a part of the lining of blood vessels, particularly in the spleen, liver, and bone marrow. These cells, called macrophages, are constituents of the reticuloendothelial system and are found in the lymph nodes, in the intestinal tract, and as free-wandering and fixed cells. As a group they have the ability to ingest not only other cells but also many other microscopic particles, including certain dyes and colloids. Within the reticuloendothelial cells erythrocytes are rapidly destroyed. Protein, including that of the hemoglobin, is broken down, and the component amino acids are transported through the plasma to be used in the synthesis of new proteins. The iron removed from hemoglobin passes back into the plasma and is transported to the bone marrow, where it may be used in the synthesis of hemoglobin in newly forming red cells. Iron not necessary for this purpose is stored within the reticuloendothelial cells but is available for release and reuse whenever it is required. In the breakdown of red cells there is no loss to the body of either protein or iron, virtually all of which is conserved and reused. In contrast, the porphyrin ring structure of hemoglobin, to which iron was attached, undergoes a chemical change that enables its excretion from the body.

Bilirubin

This reaction converts porphyrin, a red pigment, into bilirubin, a yellow pigment. Bilirubin released from reticuloendothelial cells after the destruction of erythrocytes is conveyed through the plasma to the liver, where it undergoes further changes that prepare it for secretion into the bile. The amount of bilirubin produced and secreted into the bile is determined by the amount of hemoglobin destroyed. When the rate of red cell destruction exceeds the liver's capacity to handle bilirubin, the yellow pigment accumulates in the blood, causing jaundice. Jaundice can also occur if the liver is diseased (*e.g.*, hepatitis) or if the egress of bile is blocked (*e.g.*, by a gallstone).

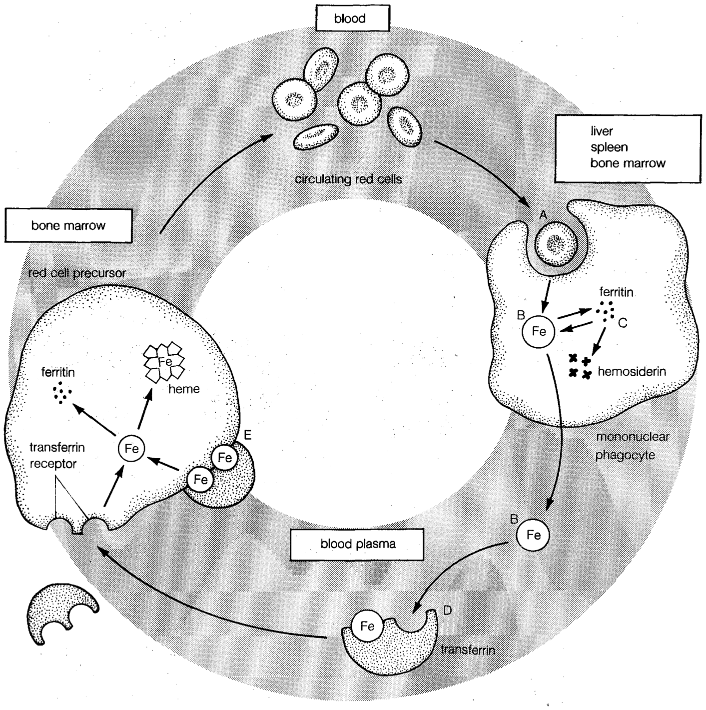

Figure 3: *The iron cycle.*
(A) Aging red cells phagocytized by mononuclear phagocytes in the liver, spleen, or bone marrow. (B) Released iron from degraded hemoglobin. (C) Excess iron stored as ferritin (a quick releaser of iron) or hemosiderin (a slow releaser). (D) The plasma protein transferrin binds iron and transports it to the bone marrow. (E) The transferrin-iron molecule enters the red cell precursor by attaching to a receptor for transferrin. Iron in food is also absorbed from the intestines and carried by transferrin to developing red cells in the bone marrow.

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White blood cells (leukocytes). White cells, unlike red cells, are nucleated and independently motile (Figures 4 and 5). Highly differentiated for their specialized functions, they do not undergo mitosis (ordinary cell division) in the bloodstream, but some retain the capability of cell division. As a group they are involved in the body's defense mechanisms and reparative activity. The number of leukocytes in normal blood ranges between 4,500 and 11,000 per cubic millimetre. Fluctuations occur during the day; lower values are obtained during rest and higher values during exercise. Violent physical exertion may cause the count to exceed 20,000 per cubic millimetre. Most of the leukocytes are outside the circulation, and the few in the bloodstream are in transit from one site to another. As living cells, their survival depends on their continuous

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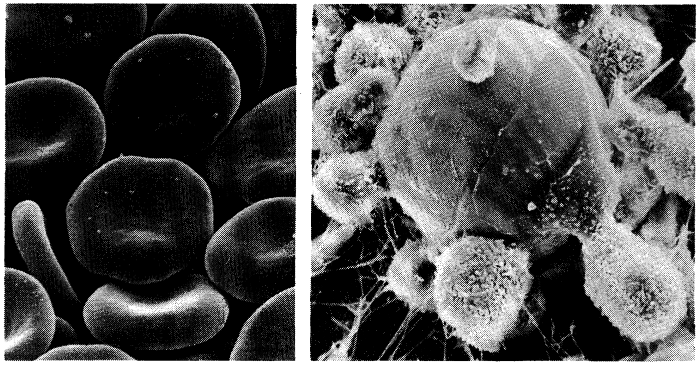

Figure 4: Electron microscope views of (left) red cells (note the biconcave silhouettes) and (right) foreign matter (in centre) being attacked and destroyed by white cells, which have extended their pseudopodia to the outer membrane of the invading cell.

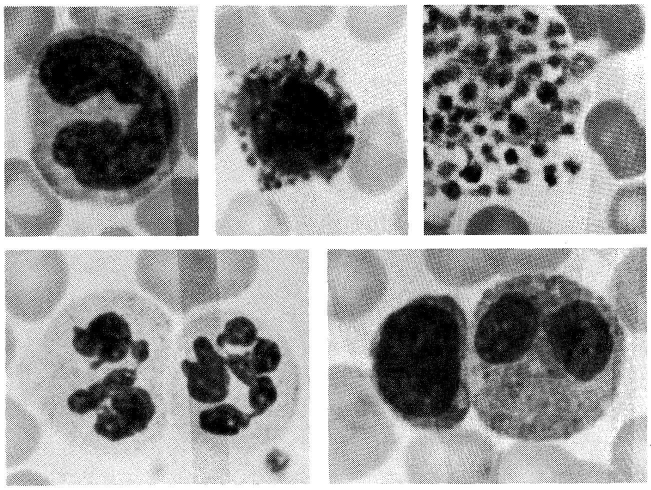

Figure 5: *White blood cells in a field of red cells.*
(Top left) Monocyte, (top centre) basophil, (top right) platelets, (bottom left) two neutrophils, (bottom right) lymphocyte and eosinophil, respectively.
A. Owczarzak/Taurus Photos, Inc.

production of energy. The chemical pathways utilized are more complex than those of the red cells and are similar to those of other tissue cells. Leukocytes, containing a nucleus and able to produce RNA, can synthesize protein. They comprise three classes of cells, each unique as to structure and function, that are designated granulocytes, monocytes, and lymphocytes.

Granulocytes. Granulocytes, the most numerous of the white cells, are larger than red cells (approximately 12–15 micrometres). They have a multilobed nucleus and contain large numbers of cytoplasmic granules (*i.e.,* granules in the cell substance outside the nucleus). Granulocytes are important mediators of the inflammatory response. There are three types of granulocytes: neutrophils, eosinophils, and basophils. Each type of granulocyte is identified by the colour of the granules when the cells are stained with a compound dye. The granules of the neutrophil are pink, those of the eosinophil are red, and those of the basophil are blue-black. About 50 to 80 percent of the white cells are neutrophils, while the eosinophils and basophils together constitute no more than 3 percent.

Neutrophils

The neutrophils are fairly uniform in size with a diameter between 12 and 15 micrometres. The nucleus consists of two to five lobes joined together by hairlike filaments. Neutrophils move with amoeboid motion. They extend long projections called pseudopods into which their granules flow; this action is followed by contraction of filaments based in the cytoplasm, which draws the nucleus and rear of the cell forward. In this way neutrophils rapidly advance along a surface. The bone marrow of a normal adult produces about 100,000,000,000 neutrophils daily (Figure 6). It takes about one week to form a mature neutrophil from a precursor cell in the marrow, yet once in the blood the mature cells live only a few hours, or perhaps a little longer after migrating to the tissues. To guard against rapid depletion of the short-lived neutrophil (for example, during infection), the bone marrow holds a large number of them in reserve to be mobilized in response to inflammation or infection. Within the body the neutrophils migrate to areas of infection or tissue injury. The force of attraction that determines the direction in which neutrophils will move is known as chemotaxis and is attributed to substances liberated at sites of tissue damage. Of the 100,000,000,000 neutrophils circulating outside the bone marrow, half are in the tissues and half are in the blood vessels; of those in the blood vessels, half are within the mainstream of rapidly circulating blood and the other half move slowly along the inner walls of the blood vessels ("marginal pool"), ready to enter tissues on receiving a chemotactic signal from them.

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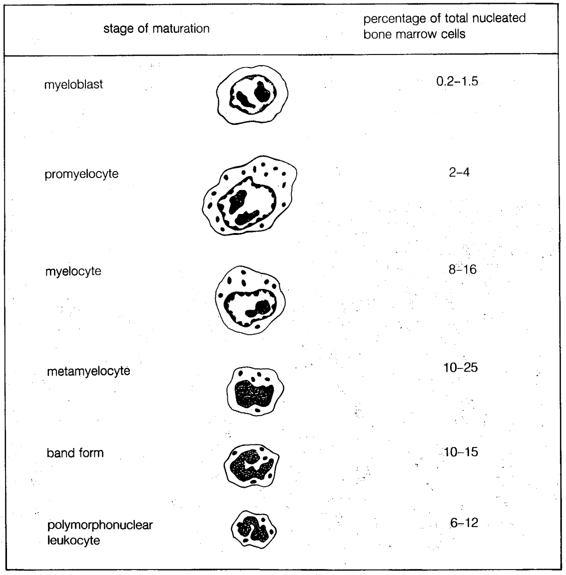

stage of maturation	percentage of total nucleated bone marrow cells
myeloblast	0.2–1.5
promyelocyte	2–4
myelocyte	8–16
metamyelocyte	10–25
band form	10–15
polymorphonuclear leukocyte	6–12

Figure 6: *Maturation of neutrophils.*
Mitotic and postmitotic stages of the neutrophil, with progressive condensation and segmentation of the nucleus after the myelocyte stage.

Neutrophils are actively phagocytic; they engulf bacteria and other microorganisms and microscopic particles. The granules of the neutrophil are microscopic packets of potent enzymes capable of digesting many types of cellular materials. When a bacterium is engulfed by a neutrophil, it is encased in a vacuole lined by the invaginated membrane. The granules discharge their contents into the vacuole containing the organism. As this occurs, the granules of the neutrophil are depleted (degranulation). A metabolic process within the granules produces hydrogen peroxide and a highly active form of oxygen (superoxide), which destroy the ingested bacteria. Final digestion of the invading organism is accomplished by enzymes.

Eosinophils and basophils

Eosinophils, like other granulocytes, are produced in the bone marrow until they are released into the circulation. Although about the same size as neutrophils, the eosinophil contains larger granules and the chromatin is generally only concentrated in two nonsegmented lobes. Eosinophils leave the circulation within hours of release from the marrow and migrate into the tissues (usually those of the skin, lung, and respiratory tract) through the lymphatic channels. Like neutrophils, eosinophils respond to chemotactic signals released at the site of cell destruction. They are actively motile and phagocytic. Eosinophils are involved in defense against parasites, and they participate in hypersensitivity and inflammatory reactions, primarily by dampening their destructive effects.

Basophils are the least numerous of the granulocytes, and their large granules almost completely obscure the underlying, double-lobed nucleus. Within hours of their release from the bone marrow, basophils migrate from the circulation to the barrier tissues (*e.g.,* the skin and mucosa), where they synthesize and store histamine, a natural modulator of the inflammatory response. When aggravated, basophils release, along with histamine and other substances, leukotrienes, which cause bronchoconstriction during anaphylaxis (a hypersensitivity reaction). Basophils incite immediate hypersensitivity reactions in association with platelets, macrophages, and neutrophils.

The largest blood cell

Monocytes. Monocytes are the largest cells of the blood (averaging 15–18 micrometres), and they make up on the average about 7 percent of the leukocytes. The nucleus is relatively big and tends to be indented or folded rather than multilobed. The cytoplasm contains large numbers of fine granules, which often appear to be more numerous near the cell membrane. Monocytes are actively motile and phagocytic. They are capable of ingesting infectious agents as well as red cells and other large particles, but they cannot replace the function of the neutrophils in the removal and destruction of bacteria. Monocytes usually enter areas of inflamed tissue later than the granulocytes. Often they are found at sites of chronic infections.

In the bone marrow, granulocytes and monocytes arise from a common precursor under the influence of the granulocyte–macrophage colony-stimulating factor (see above *Formation of blood cells*). Monocytes leave the bone marrow and circulate in the blood. After a period of hours the monocytes enter the tissues, where they develop into macrophages, the tissue phagocytes that constitute the reticuloendothelial system (or macrophage system). Macrophages occur in almost all tissues of the body: those in the liver are called Kupffer cells; those in the skin are called Langerhans cells. Apart from their role as scavengers, macrophages play a key role in immunity by ingesting antigens and processing them so that they can be recognized as foreign substances by lymphocytes.

Lymphocytes. Lymphocytes constitute about 28–42 percent of the white cells of the blood, and they are part of the immune response to foreign substances in the body. Most lymphocytes are small, only slightly larger than erythrocytes, with a nucleus that occupies most of the cell. Some are larger and have more abundant cytoplasm that contains a few granules. Lymphocytes are sluggishly motile, and their paths of migration outside of the bloodstream are different from those of granulocytes and monocytes. Lymphocytes are found in large numbers in the lymph nodes, spleen, thymus, tonsils, and lymphoid tissue of the gastrointestinal tract. They enter the circulation through lymphatic channels that drain principally into the thoracic lymph duct, which has a connection with the venous system. Unlike other blood cells, some lymphocytes may leave and reenter the circulation, surviving for about a year or more. The principal paths of recirculating lymphocytes are through the spleen or lymph nodes. Lymphocytes freely leave the blood to enter lymphoid tissue, passing barriers that prevent the passage of other blood cells. When stimulated by antigen and certain other agents, some lymphocytes are activated and become capable of cell division.

Functions and types of lymphocytes

The lymphocytes regulate or participate in the acquired immunity to foreign cells and antigens. They are responsible for immunologic reactions to invading organisms, foreign cells such as those of a transplanted organ and foreign proteins and other antigens not necessarily derived from living cells. The two classes of lymphocytes are not distinguished by the usual microscopic examination but rather by the type of immune response they elicit. The B lymphocytes (or B cells) are involved in what is called humoral immunity. Upon encountering a foreign substance (or antigen), the B lymphocyte differentiates into a plasma cell, which secretes immunoglobulin (antibodies). The second class of lymphocytes, the T lymphocytes (or T cells), are involved in regulating the antibody-forming function of B lymphocytes as well as in directly attacking foreign antigens. T lymphocytes participate in what is called the cell-mediated immune response. T lymphocytes also participate in the rejection of transplanted tissues and in certain types of allergic reactions.

All lymphocytes begin their development in the bone marrow. The B lymphocytes mature partly in the bone marrow until they are released into the circulation. Further differentiation of B lymphocytes occurs in lymphoid tissues (spleen or lymph nodes), most notably on stimulation by a foreign antigen. In humans the precursors of the T lymphocytes migrate from the marrow to the thymus, where they differentiate under the influence of a hormonelike substance. (The thymus is a small organ lying just behind the breastbone in the upper portion of the chest. It is relatively large at birth, begins to regress after puberty, and may be represented only by a fibrous cord in the elderly. The thymus begins to exert its effects on the differentiation of lymphocytes before birth. The removal of the thymus from certain animals at birth prevents the normal development of immunologic responses.) Once they have matured, the T lymphocytes leave the thymus and circulate through the blood to the lymph nodes and the spleen. The two classes of lymphocytes originally derived their names from investigations in birds, in which it was found that differentiation of one class of lymphocyte was influenced by the bursa of Fabricius (an outpouching of the gastrointestinal tract) and thus was called the B lymphocytes, and the other was influenced by the thymus and was called the T lymphocytes.

Major histocompatibility antigens

A primary function of lymphocytes is to protect the body from foreign microbes. This essential task is carried out by both T lymphocytes and B lymphocytes, which often act in concert. The T lymphocytes can only recognize and respond to antigens that appear on cell membranes in association with other molecules termed major histocompatibility complex antigens. The latter are glycoproteins that present the antigen in a form that can be recognized by T lymphocytes. In effect, T lymphocytes are responsible for continuous surveillance of cell surfaces for the presence of foreign antigens. By contrast, the antibodies produced by B lymphocytes are not confined to recognizing antigens on cell membranes; they can bind to soluble antigens in the blood or extravascular fluids. T lymphocytes typically recognize antigens of infectious organisms that must penetrate cells in order to multiply, such as viruses. During their intracellular life cycle, viruses produce antigens that appear on the cell membrane. Two classes of T lymphocytes can be involved in the response to those cell-associated viral antigens: cytotoxic T lymphocytes, which destroy the cells by a lytic mechanism; and helper T lymphocytes, which assist B cells to produce antibodies against the microbial antigens. Helper T lymphocytes exert their influence on B lymphocytes through several hormone-like peptides termed interleukins (IL). Five different T lymphocyte interleukins (IL-2, IL-3, IL-4, IL-5, and IL-6) have been discovered, each with different (and sometimes overlapping) effects on B lymphocytes and other blood cells. Interleukin-1, produced by macrophages, is a peptide that stimulates T lymphocytes and that also acts on the hypothalamus in the brain to produce fever. The ability to develop an immune response (*i.e.,* the T cell-mediated and humoral immune responses) to foreign substances is called immunologic competence (immunocompetence). Immunologic competence, which begins to develop during embryonic life, is incomplete at the time of birth but is fully established soon after birth. If an antigen is introduced into the body before immunologic competence has been established, an immune response will not result upon reinfection, and that person is said to be tolerant to that antigen.

Study of immunologic competence and immune tolerance has been accelerated by interest in organ transplantation. The success rates of organ transplantations have been improved by better knowledge about donor selection and improved techniques for suppressing the immune responses of the recipient. An important element in donor selection is tissue typing: the matching of the donor's histocompatibility antigens (human leukocyte antigens) with those of the prospective recipient. The closer the match, the greater the probability that the graft will be accepted. (See also TRANSPLANTS, ORGAN AND TISSUE.)

Platelets (thrombocytes). The blood platelets are the smallest cells of the blood, averaging about two to four micrometres in diameter. Although much more numerous (150,000 to 400,000 per cubic millimetre) than the white cells, they occupy a much smaller fraction of the volume of the blood because of their relatively minute size. Like the red cells they lack a nucleus and are incapable of cell division, but they have a more complex metabolism and internal structure than have the red cells. When seen in fresh blood they appear spheroid, but they have a tendency to extrude hairlike filaments from their membranes. They adhere to each other but not to red cells and white cells. Tiny granules within platelets contain substances important for the clot-promoting activity of platelets.

Platelet metabolism

The function of the platelets is related to hemostasis, the prevention and control of bleeding. When the endothelial surface (lining) of a blood vessel is injured, platelets in large numbers immediately attach to the injured surface and to each other, forming a tenaciously adherent mass of platelets. The effect of the platelet response is to stop the bleeding and to form the site of the developing blood clot, or thrombus. If platelets are absent, this important defense reaction cannot occur, and protracted bleeding from small wounds (prolonged bleeding time) results. The normal resistance of capillary membranes to leakage of red cells is dependent upon platelets. Severe deficiency of platelets reduces the resistance of the capillary walls, and abnormal bleeding from the capillaries occurs, either spontaneously or as the result of minor injury. Platelets also contribute substances essential for the normal coagulation of the blood, and they cause the shrinking, or retraction, of a clot after it has been formed. (For additional information about the causes, consequences, and treatment of thrombosis, see below *Bleeding and blood clotting.*)

Platelets are formed in the bone marrow by segmentation of the cytoplasm (the cell substance other than the nucleus) of cells known as megakaryocytes, the largest cells of the marrow. Within the marrow the abundant granular cytoplasm of the megakaryocyte divides into many small segments that break off and are released as platelets into the circulating blood. After about 10 days in the circulation, platelets are removed and destroyed. There are no reserve stores of platelets except in the spleen, in which platelets occur in higher concentration than in the peripheral blood. Some platelets are consumed in exerting their hemostatic effects, and others, reaching the end of their life span, are removed by reticuloendothelial cells (any of the tissue phagocytes). The rate of platelet production is controlled but not so precisely as the control of red cell production. A hormonelike substance, thrombopoietin, which has not been identified chemically, is believed to be the chemical mediator that regulates the number of platelets in the blood by stimulating an increase in the number and growth of megakaryocytes, thus controlling the rate of platelet production.

EXAMINATION OF THE BLOOD IN THE LABORATORY

Physicians rely upon laboratory analysis to obtain measurements of many constituents of the blood, information useful or necessary for the detection and recognition of disease. Hemoglobin contains a highly coloured pigment that interferes with the passage of a beam of light. To measure hemoglobin concentration, blood is accurately diluted and the red cells broken down to yield a clear red solution. A photoelectric instrument is used to measure the absorbance of transmitted light, from which hemoglobin concentration can be calculated. Changes in the hemoglobin concentration of the blood are not necessarily directly paralleled by changes in the red cell count and the hematocrit value, because the size and hemoglobin concentration of red cells may change in disease. Therefore, measurements of the red cell count and the hematocrit value may provide useful information as well. These three tests can be carried out rapidly and in large numbers by automated machines. Electronic particle counters for determining red cell, white cell, and platelet counts are widely used. Only a drop of blood is needed for the analyses, which are completed within a minute; results are printed in a laboratory report that is sent to the physician. Although expensive, the equipment increases the output of the laboratory and saves the technician valuable time.

Blood tests

Adequate examination of the blood cells requires that a thin film of blood be spread on a glass slide, stained with a special blood stain (Wright's stain), and examined under the microscope. Individual red cells, white cells, and platelets are examined, and the relative proportions of the several classes of white cells are tabulated. The results may have important diagnostic implications. In iron deficiency anemia, for example, the red cells look paler than normal because they lack the normal amount of hemoglobin; in malaria the diagnosis is established by observing the malarial parasites within the red cells. In pneumonia and many infections the proportion of neutrophilic leukocytes is usually increased, while in others, such as whooping cough and measles, there is an increase in the proportion of lymphocytes.

Chemical analyses

Chemical analyses measure many of the constituents of plasma. Often serum rather than plasma is used, however, since serum can be obtained from clotted blood without the addition of an anticoagulant. Changes in the concentrations of chemical constituents of the blood can indicate the presence of a disease process. For example, quantitative determination of the amount of sugar in the blood is essential for the diagnosis of diabetes, a disease in which the blood sugar tends to be elevated. Nitrogenous waste products, in particular urea, tend to accumulate in patients with diseased kidneys that are unable to excrete these substances at a normal rate. An increase in the concentration of bilirubin in the serum often reflects a disorder of the liver and bile ducts or an increased rate of destruction of hemoglobin. Measurements of these and many other serum constituents are so valuable in medical diagnosis that often multiple tests are performed.

Tests can be performed manually using an individual procedure for each analysis; however, the autoanalyzer, a completely automated machine, increases the number of chemical analyses that can be performed in laboratories. A dozen analyses may be made simultaneously by a single machine employing a small amount of serum. The serum is automatically drawn from a test tube and is propelled through plastic tubing of small diameter. As the serum specimen advances, it is divided; appropriate reagents are added; chemical reactions occur with formation of a product that can be measured with a photoelectric instrument; and the result appears as a written tracing from which serum concentration of various substances can be read directly. The data acquired by the machine may be fed automatically into a computer and the numerical results printed on a form that is submitted to a physician. Many of the available analyses are not performed routinely but are invaluable in special circumstances. In cases of suspected lead poisoning, for example, detection of an elevated level of lead in the blood may be diagnostic. Some analytical procedures have specific diagnostic usefulness. These include assays for certain hormones, including measurement of the thyroid hormone in the serum of patients suspected of having thyroid disease.

Blood typing

Other important laboratory procedures are concerned with immunologic reactions of the blood. Careful determinations of the blood groups of the patient and of the blood donor, and cross matching of the cells of one with the serum of the other to ensure compatibility, are essential for the safe transfusion of blood. The Rh type of a pregnant woman is regularly determined and is necessary for the early detection of fetal–maternal incompatibility and for proper prevention or treatment of hemolytic disease of the newborn. The diagnosis of certain infectious diseases depends upon the demonstration of antibodies in the patient's serum.

Many other kinds of blood examination yield useful results. Enzymes normally present in the muscle of the heart may be released into the blood when the heart is damaged by a coronary occlusion (obstruction of the coronary artery) with consequent tissue necrosis. Measurement of these enzymes in the serum is regularly performed to assist in diagnosis of this type of heart disease. Damage to the liver releases other enzymes, measurement of which aids in evaluation of the nature and severity of liver disease. Inherited abnormalities of proteins are increasingly recognized and identified by use of sophisticated methods. Accurate diagnosis of hemophilia and other bleeding disorders is made possible by investigations of the coagulation mechanism. Measurements of the concentration of folic acid and vitamin B_{12} in the blood provide the basis for diagnosis of deficiencies of those vitamins. The number of potentially useful blood tests is so vast that they must be selected judiciously in the evaluation of the individual patient. (C.L.C./R.S.S.)

Blood groups

Blood groups are classifications of blood based on inherited differences (polymorphisms) in antigens on the surfaces of the red cells. Inherited differences of white cells, platelets, and plasma proteins also constitute blood groups, but they are not included in this discussion.

GENERAL SURVEY OF BLOOD GROUPING

Historical background. William Harvey, the English physician, announced his observations on the circulation of the blood in 1616 and published his famous monograph entitled *Exercitatio Anatomica de Motu Cordis et Sanguinis in Animalibus* (*The Anatomical Exercises Concerning*

The first human transfusions

the Motion of the Heart and Blood in Animals, 1653) in 1628. His discovery, that blood circulates around the body in a closed system, was an essential prerequisite of the concept of transfusing blood from one animal to another of the same or different species. In England experiments on the transfusion of blood were pioneered in dogs in 1665 by Richard Lower, an English physician. A later experiment, in which dogs were also used, is described in the diary of Samuel Pepys in the entry for Nov. 14, 1666. In November 1667 Lower transfused the blood of a lamb into a man; this event also is recorded in Pepys's diary. Meanwhile, in France, Jean-Baptiste Denis had also been transfusing lambs' blood into human subjects and described what is probably the first recorded account of the signs and symptoms of a hemolytic transfusion reaction. After a fatality Denis was arrested, and the procedure of transfusing the blood of other animals into humans was prohibited by an act of the Chamber of Deputies in 1668, unless sanctioned by the Faculty of Medicine of Paris. Ten years later, in 1678, the British Parliament also prohibited transfusions. Little advance was made in the next 150 years.

In 19th-century England interest was reawakened by the activities of James Blundell, whose humanitarian instincts had been aroused by the frequently fatal outcome of hemorrhage occurring after childbirth. He insisted that it was better to use human blood for transfusion in such cases. During the years 1875–1900 another hiatus of interest resulted from the introduction of the treatment of shock by injecting physiological saline solution (salt solution of the same osmotic pressure as the blood).

Discovery of ABO blood groups

In 1875 the German physiologist Leonard Landois published a monograph reviewing 347 transfusions between humans and 129 transfusions between animals. He showed that, if the red blood cells of an animal belonging to one species were mixed with serum taken from an animal of another species, the red cells usually clumped and that sometimes the red cells burst—*i.e.,* hemolyzed. He attributed the appearance of black urine after transfusion of heterologous blood (blood from a different species) to the hemolysis of the incompatible red cells. Thus, the dangers of transfusing blood of another species to humans were established scientifically. Safe transfusion rests primarily on two more advances, the first of which was the discovery of the human ABO blood groups by the Austrian-born American biologist Karl Landsteiner in 1901. Landsteiner found that there were substances in the blood, antigens and antibodies, that induced clumping of red cells when red cells of one type were added to those of a second type. He recognized three groups—A, B, and O—based on their reaction to each other. A fourth group, AB, was identified a year later by another research team. The red cells of an individual contain antigens on their surfaces that correspond to their blood group and antibodies in the serum that identify and combine with the antigen sites on the surfaces of red cells of another type. Red cells of the A group clump with donor blood of the B group; those of the B group with blood of the A group; those of the AB group with those of the A or the B group because AB cells contain both A and B antigens; and those of the O group do not generally clump with any group because they do not contain either A or B antigens. The ABO blood groups and other groups discovered later, such as Kell, Diego, Lutheran, Duffy, and Kidd, are determined in this manner, and it is in this way that blood groups are defined. In 1939–40 the rhesus (Rh) blood group system was discovered as a result of similar techniques.

The second advance involved the use of substances such as sodium citrate that prevent the clotting of blood being used in transfusion. These substances are innocuous when given by infusion to humans. In 1914 Luis Agote demonstrated that small, nontoxic quantities of sodium citrate could prevent the coagulation of blood. The introduction of citrate transformed the hectic, nerve-racking procedure of direct transfusion into a simple, relaxed operation and eventually led to the development of blood storage. In 1943 John Freeman Loutit and Patrick Loudon Mollison acidified the citrate and added dextrose, which allowed storage of red cells at 4° C for 21 days. Minor modifications in the formula and the addition of adenine have further increased storage time of blood in the liquid state. Addition of a cryoprotective agent (glycerol) makes frozen storage of red cells possible. The first blood banks were opened in the United States in 1937.

Definitions of antigen and antibody

The importance of antigens and antibodies. As has been stated previously, blood groups are determined by the presence of antigens on the surfaces of the red cells. An antigen has been defined as a substance that can, in certain circumstances, induce the production of the corresponding antibody. (An antigen that can induce antibody production is also called an immunogen.) An antibody is a substance that is capable of reacting with a specific antigen to counter its effects.

The human red cells carry many antigens. Because the reaction between red cells and corresponding antibodies usually results in clumping—agglutination—of the red cells, the antigens on the surfaces of these red cells are often referred to as agglutinogens. Antibodies are part of the circulating plasma proteins known as immunoglobulins. Immunoglobulins are classified by molecular size and weight and by several other biochemical properties. Most blood group antibodies are found either on immunoglobulin G (IgG) or immunoglobulin M (IgM) molecules, but occasionally the immunoglobulin A (IgA) class may exhibit blood group specificity.

It has been customary to distinguish between naturally occurring antibodies and immune antibodies. So-called naturally occurring antibodies are really the result of immunization by substances in nature that have structures similar to human blood groups. Naturally occurring antibodies are present in an individual despite the fact that there has been no previous exposure to the corresponding red cell antigens—for example, anti-A in the plasma of people of blood group B and anti-B in the plasma of people of blood group A. Immune antibodies—for example, Rh antibodies (see below)—occur only if they are invoked by exposure to the corresponding red cell antigens. Immunization (*i.e.,* the production of antibodies in response to antigen) against blood group antigens in humans can occur as a result of pregnancy (see below *Blood groups and disease*), therapeutic transfusion, or deliberate immunization. The combination of pregnancy and transfusion is a particularly potent stimulus. Individual blood group antigens vary in their antigenic potential; for example, some of the antigens belonging to the Rh and ABO systems are strongly immunogenic (*i.e.,* capable of inducing antibody formation), whereas the antigens of the Kidd and Duffy blood group systems are much weaker immunogens.

The MN, P, and Rh systems

Discovery of the individual blood groups. In Table 1 the well-established human blood group systems are listed in the order of discovery. The discovery of the ABO blood group system in 1901 was the result of a deliberate search for intraspecies differences among humans. The application of knowledge of the ABO system in blood transfusion practice is of enormous importance, since mistakes can have fatal consequences. Twenty-six years after the initial discovery, Landsteiner, with his colleague Philip Levine, discovered the MN (later called MNSs) and P blood group systems because they had the idea of immunizing another species with human blood. The discovery of the Rh sys-

Table 1: Major Human Blood Group Systems

system	date of discovery	main antigens
ABO	1901	A_1, A_2, B, H
MNSs	1927	M, N, S, s
P	1927	P_1, P_2
Rh	1940	D, C, c, E, e
Lutheran	1945	Lu^a, Lu^b
Kell	1946	K, k
Lewis	1946	Le^a, Le^b
Duffy	1950	Fy^a, Fy^b
Kidd	1951	Jk^a, Jk^b
Diego	1955	Di^a, Di^b
Yt	1956	Yt^a, Yt^b
I	1956	I, i
Xg	1962	Xg^a
Dombrock	1965	Do^a

tem by Landsteiner and Alexander S. Wiener in 1940 was made because they tested human red cells with antisera developed in rabbits and guinea pigs by immunization of the animals with the red cells of the rhesus monkey *Macaca mulatta.* A few months elapsed before it was realized that antibodies induced by antigens of the Rh system could be found in human beings as a result of pregnancies or blood transfusions.

The remaining blood group systems were first described after antibodies were identified in patients. Frequently, such discoveries resulted from the search for the explanation of an unexpected unfavourable reaction in a recipient after a transfusion with formerly compatible blood. In such cases the antibodies in the recipient were produced against previously unidentified antigens in the donor's blood. In the case of the Rh system, for example, the presence of antibodies in the maternal serum directed against antigens present on the child's red cells can have serious consequences because of antigen–antibody reactions that produce hemolytic disease of the newborn, or erythroblastosis fetalis. Some of the other blood group systems—for example, the Kell and Kidd systems—were discovered because an infant was found to be suffering from hemolytic disease of the newborn even though mother and child were compatible as far as the Rh system was concerned.

Presence of ABO antigens in tissues

Blood group antigens in tissues. The blood group antigens are inherited characteristics; they are present on the red cells as a property of the cell membrane, although they are not restricted solely to red cells or even to hematopoietic tissues. There is no doubt that the antigens of the ABO system are widely distributed throughout the tissues. These antigens have been unequivocally identified on platelets and white cells (both lymphocytes and polymorphonuclear leukocytes) and in skin, the epithelial (lining) cells of the gastrointestinal tract, the kidney, the urinary tract, and the lining of the blood vessels. Evidence for the presence of the antigens of other blood group systems on cells other than red cells is less well substantiated, and sometimes the opinions of different workers disagree. Among the red cell antigens, only those of the ABO system are regarded as tissue antigens and therefore need to be considered in transplantation.

As they occur on the red cells, the blood group substances are in an alcohol-soluble form. In the ABO and Lewis systems the blood group specific substances also occur in tissue fluids and secretions in a water-soluble form. In the case of the Lewis system the antigens are present primarily in the secretions, and their presence on the red cells is a consequence of passive adsorption onto the cells from the plasma.

Chemistry of the blood group substances. The exact chemical structure of some blood groups has been identified, as have the gene products (*i.e.,* those molecules synthesized as a result of an inherited genetic code on a gene of a chromosome) that assist in synthesizing the antigens on the red cell surface which determine the blood type. Blood group antigens are present on glycolipid and glycoprotein molecules of the red cell membrane. The carbohydrate chains of the membrane glycolipids are oriented toward the external surface of the red cell membrane and carry antigens of the ABO, Hh, Ii, and P systems. Glycoproteins, which traverse the red cell membrane, have a polypeptide backbone to which carbohydrates are attached. An abundant glycoprotein, band 3, contains ABO, Hh, and Ii antigens. Another integral membrane glycoprotein, glycophorin A, contains large numbers of sialic acid molecules and MN blood group structures; another, glycophorin B, contains Ss and U antigens.

Effects of glycosyltransferase

The genes responsible for inheritance of ABH and Lewis antigens are glycosyltransferases (a group of enzymes that catalyzes the addition of specific sugar residues to the core precursor substance). For example, the *H* gene codes for the production of a specific glycosyltransferase that adds L-fucose to a core precursor substance, resulting in the H antigen; the *Le* gene codes for the production of a specific glycosyltransferase that adds L-fucose to the same core precursor substance, but in a different place, forming the Lewis antigen; the *A* gene adds *N*-acetyl-D-galactosamine (H must be present), forming the A antigen; and the *B* gene adds D-galactose (H must be present), forming the B antigen. The P system is analogous to the ABH and Lewis blood groups in the sense that the P antigens are built by the addition of sugars to precursor globoside and paragloboside glycolipids, and the genes responsible for these antigens must produce glycosyltransferase enzymes.

The genes that code for MNSs glycoproteins change two amino acids in the sequence of the glycoprotein to account for different antigen specificities. Additional analysis of red cell membrane glycoproteins have shown that in some cases the absence of blood group antigens is associated with an absence of minor membrane glycoproteins that are present normally in antigen-positive persons.

METHODS OF BLOOD GROUPING

Identification of blood groups. The basic technique in identification of the antigens and antibodies of blood groups is the agglutination test. Agglutination of red cells results from antibody cross-linkages established when different specific combining sites of one antibody react with antigen on two different red cells. By mixing red cells (antigen) and serum (antibody), either the type of antigen or the type of antibody can be determined depending on whether a cell of known antigen composition or a serum with known antibody specificity is used.

In its simplest form, a volume of serum containing antibody is added to a thin suspension (2–5 percent) of the red cells suspended in physiological saline solution in a small tube with a narrow diameter. After incubation for an adequate time, at the appropriate temperature, the red cells will have settled to the bottom of the tube. These sedimented red cells are examined macroscopically (with the naked eye) for agglutination, or they may be spread on a slide and viewed through a low-power microscope.

Complete and incomplete antibody

An antibody that agglutinates red cells when they are suspended in saline solution is called a complete antibody. With powerful complete antibodies, such as anti-A and anti-B, agglutination reactions visible to the naked eye take place when a drop of antibody is placed on a tile together with a drop containing red cells in suspension. After stirring, the tile is rocked, and agglutination is visible in a few moments. It is always necessary in blood grouping to include a positive and negative control for each test that is done.

An antibody that does not clump red cells when they are suspended in saline solution is called incomplete. Such antibodies block the antigenic sites of the red cells so that subsequent addition of complete antibody of the same antigenic specificity does not result in agglutination. Incomplete antibodies will agglutinate red cells carrying the appropriate antigen, however, when the cells are suspended in media containing protein. Albumin from the blood of cattle is a substance that is frequently used for this purpose. Red cells may also be rendered specifically agglutinable by incomplete antibodies after treatment with such protease enzymes as trypsin, papain, ficin, or bromelain.

The Coombs test. When an incomplete antibody reacts with the red cells in saline solution, the antigenic sites become coated with antibody globulin (gamma globulin) and no visible agglutination reaction takes place. The presence of antibody globulin on cells can be detected by a test that is called the Coombs test, after its inventor, the English immunologist Robert Coombs. The Coombs reagent (also called antiglobulin) is made by immunizing rabbits with human immunoglobulin. The rabbits respond by making antihuman globulin (*i.e.,* antibodies against human immunoglobulin and complement) that must be purified before use. The antiglobulin reagent or solution usually contains antibodies against IgG and complement. The Coombs reagent is added to the washed cells, the tube is centrifuged, and if they are coated by antibody globulin or complement, agglutinates will form. Newer antiglobulin reagents (made by immunizing with purified protein) can detect either globulin or complement. Depending on how it is performed, the Coombs test can detect incomplete antibody in the serum or antibody bound to the red cell membrane. In certain diseases, anemia may be caused by the coating of red cells with antibody globulin. This can happen when a mother has made antibodies against the red cells of her newborn child or if a person makes an autoantibody against his own red cells.

Adsorption, elution, and titration. If a serum contains a mixture of antibodies, it is possible to prepare pure samples of each by a technique that is called adsorption. By this technique an unwanted antibody is removed by mixing it with red cells carrying the appropriate antigen. The antigen interacts with the antibody and binds it to the cell surface. These red cells are washed thoroughly and spun down tightly by centrifugation, all the fluid above the cells is removed, and the cells are then said to be packed. The cells are packed to avoid dilution of the antibody being prepared. Adsorption, then, is a method of separating mixtures of antibodies by removing some and leaving others. It is used to identify antibody mixtures and to purify reagents. The purification of the Coombs reagent, mentioned earlier, is done in the same way.

If red cells have adsorbed antibody globulin onto their surfaces, the antibody can sometimes be recovered by a process known as elution. One simple way of eluting (dissociating) antibody from washed red cells is to heat them at 56° C (133° F) in a small volume of saline solution. Other methods include use of acid or ether. This technique is sometimes useful in the identification of antibodies.

Titration technique

The technique used to determine the strength of an antibody is called titration. Doubling dilutions of the antibody are made in a suitable medium in a series of tubes. Cells carrying the appropriate antigen are added, and the agglutination reactions are read and scored for the degree of positivity. The actual titre (concentration) of the antibody is given by the dilution at which some degree of agglutination, however weak, can still be seen. This would not be a safe dilution to use for blood-grouping purposes. If an antiserum can be diluted, the dilution chosen must be such that strong positive reactions occur with selected positive control cells. Titration is helpful when preparing reagents and comparing antibody concentrations at different time intervals.

Inhibition tests. Inhibition tests are used to detect the presence of antigen with blood group specificity in solutions; inhibition of a known antibody–antigen reaction by a fluid indicates a particular blood group specificity. If an active substance is added to antibody, neutralization of the antibody's activity prevents agglutination when red cells carrying the appropriate antigen are subsequently added to the mixture. A, B, Lewis, Chido, Rogers, and P antigens are readily available and can be used to facilitate antibody identification. This technique was used to elucidate the biochemistry of ABH, Ii, and Lewis systems, and it is important in forensic medicine as a means of identifying antigens in blood stains. Chido and Rogers are present in plasma. P substance has been isolated from cyst fluid. ABH and Lewis substances (appropriate to their blood type) are also present in saliva of secretor (*Se*) individuals.

Hemolysis. Laboratory tests in which hemolysis (destruction) of the red cells is the end point are not used frequently in blood grouping. For hemolysis to take place, a particular component of fresh serum called complement must be present. Complement must be added to the mixture of antibody and red cells. It may sometimes be desirable to look for hemolysins that destroy group A red cells in mothers whose group A children are incompatible or in individuals, not belonging to groups A or AB, who have been immunized with tetanus toxoid that contains substances with group A specificity.

Hemolytic reactions may occur in patients who have been given transfusions of blood that either is incompatible or has already hemolyzed. The sera of such patients require special investigations to detect the presence of hemoglobin that has escaped from red cells destroyed within the body and for the breakdown products of other red cell constituents.

Naturally occurring antibodies

Sources of antibodies and antigens. Normal donors are used as the source of supply of the so-called naturally occurring antibodies, such as those of the ABO, P, and Lewis systems. These antibodies work best at temperatures below that of the body (37° C, or 98.6° F); in the case of what are known as cold agglutinins, such as anti-P$_1$, the antibody is most active at 4° C (39° F). Most antibodies used in blood grouping must be searched for in immunized donors.

Antibodies for MN typing are usually raised in rabbits—similarly for the Coombs reagent. Antibodies prepared in this way have to be absorbed free of unwanted components and carefully standardized before use. Additional substances with specific blood group activity have been found in certain plants. Plant agglutinins are called lectins. Some useful reagents extracted from seeds are anti-H from *Ulex europaeus* (common gorse); anti-A$_1$, from another member of the pulse family Fabaceae (Leguminosae), *Dolichos biflorus;* and anti-N from a South American plant, *Vicia graminea.* Agglutinins have also been found in animals; for example, the fluid pressed from the land snail *Octala lactea.* Additional plant lectins and agglutinins from animal fluids have been isolated.

Monoclonal antibodies (structurally identical antibodies produced by hybridomas) to blood groups are replacing some of the human blood grouping reagents. Mouse hybridomas (hybrid cells of a myeloma tumour cell and lymphocyte merging) produce anti-A and anti-B monoclonal antibodies. The antibodies are made by immunizing with either red cells or synthetic carbohydrates. In addition to their use in blood grouping, these monoclonal antibodies can be of use in defining the hereditary background (heterogenicity) and structure of the red cell antigen.

DIFFICULTIES IN BLOOD GROUPING

Effects of disease on cells

Suspensions of cells suitable for blood grouping can be prepared from blood taken with or without an anticoagulant. Serologists have individual preferences. After such infections as pneumonia, red cells may become agglutinable by almost all normal sera because of exposure of a hitherto hidden antigenic site (T), as a result of the action of bacterial enzymes. When the patient recovers, the blood also returns to normal with respect to agglutination. It is unusual for the red cells to reflect antigenicity other than that determined by the individual's genetic makeup. The presence of an acquired B antigen on the red cells has been described occasionally in diseases of the colon, thus allowing the red cell to express an antigenicity other than that genetically determined. Other diseases may alter immunoglobulins; for example, some may induce the production of antibodies directed against the person's own blood groups (autoimmune hemolytic anemia) and thus may interfere with blood grouping. In other diseases, a defect in antibody synthesis may cause the absence of anti-A and anti-B antibody.

Preservation of antibodies. Blood group antibodies are relatively stable and can be stored for extended periods of time. At 4° C the usual shelf life is two years. Frozen storage can extend the shelf life even longer. The reagents must be appropriately standardized and tested with both positive and negative controls when used for blood grouping. Because contaminated or infected sera may give incorrect results, the reagents contain sodium azide to prevent contamination. A minimal amount of protein (approximately 1 percent) may need to be added to the reagent to stabilize the antibody. Other additives and chemical modification of the antibody molecules to enhance agglutination may also be used to improve the reagents.

Identification of antibodies. When an antibody is found that apparently does not correspond to any that are already known, it must first be tested with a large panel of red cells of known antigen makeup, which should include representation from different races. (An antigen common in one population may be rare in a different racial group.) In most cases the reactivity pattern corresponds to the specificity of a known antigen or antigens. In occasional cases, no specificity is detected and the possibility exists that either a new blood group or new antigen within a blood group has been found. Statistical evaluation can then be made to see whether the reactions of the new antibody are independent of all the systems known at the time of testing. Finally, the pattern of inheritance of a new blood group antigen can be established by testing families whose members have been typed for all the known systems. This is the way in which the inheritance patterns of a new system are demonstrated.

USES OF BLOOD GROUPING

Tests of blood donors

Medical uses. *Transfusion.* The blood donated by healthy persons is tested to ensure that the level of hemoglobin is satisfactory and that there is no risk of transmitting certain diseases, such as venereal disease or liver infection. It is then fractionated (split) into its component parts, particularly red cells, plasma, and platelets. Correct matching for the ABO system is vital. Compatible donors on the basis of their possessing A, B, or O blood are shown in Table 2.

Table 2: The ABO and Rh Groups in Transfusion

system	recipient type	donor red cell type	donor plasma type
ABO	A	A* or O	A or AB
ABO	B	B or O	B or AB
ABO	O	O only	O, A, B, or AB
ABO	AB	A*B, A*, B or O	AB
Rh	positive	positive or negative	positive or negative
Rh	negative	negative or positive†, ‡	negative or positive†

*Not if the patient's serum contains anti-A$_1$ (antibody to common type A red cell in subgroup A patients). †Not if the patient is a female less than 45 years old (child-bearing possible), unless life-threatening hemorrhage is present and transfusion of Rh-positive blood is lifesaving. ‡Not if the patient's serum contains anti-D (antibody to positive red cells) except under unusual medical circumstances.

As explained above, the most important blood group systems for transfusion of red cells are ABO and Rh. Persons who have either of the red cell antigens (A and B) have antibody present in their serum of the type that will oppose an antigen of its opposite nature; for example, group A blood contains A antigens on red cell surfaces and anti-B antibodies in the surrounding serum. On the other hand O group individuals lack both the A and the B antigen and thus have both anti-A and anti-B in their serum. If these antibodies combine with the appropriate antigen, the result is hemolytic transfusion reaction, and possibly death. Red cell transfusions must therefore be ABO compatible. The blood groups A and B have various subgroups (called, for example, A$_1$, A$_2$, A$_3$, and B$_1$, B$_2$, and B$_3$, respectively). The only common subgroups that are likely to affect red cell transfusions are the subgroups of A.

Potential donors are also tested for some of the antigens of the Rh system, since it is essential to know whether they are Rh-positive or Rh-negative. Rh-negative indicates the absence of the D (Rh$_0$) antigen. Rh-negative persons transfused with Rh-positive blood will make anti-D antibodies from 50 to 75 percent of the time. Alloantibody (antibody made in response to a foreign red cell antigen) is usually not harmful but does require subsequent transfusions to be antigen-negative. Rh-positive blood should never be given to Rh-negative females before or during the childbearing age unless Rh negative blood is not available and the transfusion is livesaving. If such a woman subsequently became pregnant with an Rh-positive fetus, she might form anti-Rh antibody, even though the pregnancy was the first, and the child might suffer from hemolytic disease of the newborn.

Direct match test

Care must be taken not to give a transfusion unless the cells of the donor have been tested against the recipient's serum. If this compatibility test indicates the presence of antibodies in the recipient's serum for the antigens carried by the donor's cells, the blood is not suitable for transfusion because an unfavourable reaction might occur. The test for compatibility is called the direct match test. It involves testing the recipient's serum with the donor's cells, suspended in saline solution at 20° C and 37° C, in albumin at 37° C, and by the indirect Coombs test at 37° C. These are adequate screening tests for most naturally occurring and immune antibodies.

If, in spite of all the compatibility tests, a reaction does occur after the transfusion is given (the unfavourable reaction often manifests itself in the form of a fever), an even more careful search must be made for any red cell antibody that might be the cause. A reaction after transfusion is not necessarily due to red cell antigen–antibody reactions. It could be caused by the presence of antibodies to the donor's platelets or white cells. Transfusion reactions are a particular hazardfor persons requiring multiple transfusions.

Usefulness of ABO antigens

Organ transplants. The ABO antigens are widely distributed throughout the tissues of the body. Therefore, when organs such as kidneys are transplanted, most surgeons prefer to use organs that are matched to the recipient's with respect to the ABO antigen system, although the occasional survival of a grafted ABO-incompatible kidney has occurred. The remaining red cell antigen systems are not relevant in organ transplantation.

Blood groups and disease. The question of whether the blood groups merely have nuisance value in clinical practice has led to a search for evidence of associations with disease. In some cases an increased incidence of a particular antigen seems to be associated with a certain disease. Cancer of the stomach is more common in people of group A than in those of groups O and B. Duodenal ulceration is more common in nonsecretors of ABH substances than in secretors. For practical purposes, however, these statistical correlations are unimportant. There are other examples that illustrate the importance of blood groups to the normal functions of red cells.

Persons who lack all Rh antigens have been described. In such people red cells of altered shape (stomatocytes) and a mild compensated hemolytic anemia (Rh$_{null}$ disease) are present. The McLeod phenotype (weak Kell antigens and no K$_x$ antigen) is associated with an acanthocytosis and a compensated hemolytic anemia. Blood group antigens or related structures may also serve as receptors. There is evidence that Duffy-negative human red cells are resistant to infection by *Plasmodium knowlesi,* a simian malaria parasite. Other studies indicate that *Plasmodium falciparum* receptors may reside on glycophorin A and may be related to the Wr^b antigen. Blood group incompatibility between mother and child can cause hemolytic disease of the newborn. In this disease IgG blood group antibody molecules cross the placenta, enter the fetal circulation, react with the fetal red cells, and destroy them. Only certain blood group systems cause hemolytic disease of the newborn, and the severity of the disease in the fetus varies greatly. ABO incompatibility usually leads to mild disease. Rh, or D antigen, incompatibility is now largely preventable by treating Rh-negative mothers with Rh immunoglobulin (antibody), which prevents immunization (forming antibodies) to the D antigen. Many other Rh antigens, as well as other red cell group antigens, cause hemolytic disease of the newborn. The baby may be anemic at birth, which can be treated by transfusion with antigen-negative red cells. Even total exchange transfusion may be necessary. In some cases, transfusions may be given while the fetus is still within the uterus, so-called "intrauterine transfusion." Another common problem for the baby is bilirubin, a breakdown product of hemoglobin. Hyperbilirubinemia may lead to neurological deficits. Exchange transfusion eliminates most of the hemolysis by providing red cells, which do not react with the antibody. It also decreases the amount of antibody and allows the child to recover from the disease. Once the antibody disappears, the child's own red cells survive normally.

Medicolegal uses. Although blood group studies cannot be used to prove paternity, they can provide unequivocal evidence that a male is not the father of a particular child. Since the red cell antigens are inherited as dominant traits, a child cannot have a blood group antigen that is not present in one or both parents. For example, if the child

Table 3: Exclusions of Paternity on the ABO System

matings	possible children	impossible children
O×O	O	A, B, AB
O×A	O, A	B, AB
O×B	O, B	A, AB
O×AB	A, B	O, AB
A×A	O, A	B, AB
A×B	O, A, B, AB	
A×AB	A, B, AB	O
B×B	O, B	A, AB
B×AB	A, B, AB	O
AB×AB	A, B, AB	O

Evidence against paternity

in question belongs to group A and both the mother and the putative father are group O, the man is excluded from paternity. Table 3 shows the phenotypes (observed characters) of the offspring that can and cannot be produced in the matings on the ABO system, considering only the three alleles (alternative genes) *A, B,* and *O.* Similar inheritance patterns are seen in all blood group systems. Furthermore, if one parent is genetically homozygous for a particular antigen—that is, has inherited the gene for it from both the grandfather and grandmother of the child—then that antigen must appear in the blood of the child. For example, on the MN system, a father whose phenotype is M and whose genotype is *MM* (in other words, a father who is of blood type M and has inherited the characteristic from both parents) must transmit an *M* allele to all his progeny.

In medicolegal work it is important that the blood samples are properly identified. By using multiple red cell antigen systems and adding additional studies on other blood types (HLA [human leukocyte antigen], red cell enzymes, and plasma proteins), it is possible to state with a high degree of statistical certainty that a particular male is the father.

GENETIC AND EVOLUTIONARY SIGNIFICANCE OF BLOOD GROUPS

Blood groups as genetic markers

Blood groups and genetic linkage. Red cell groups act as markers (inherited characteristics) for genes present on chromosomes which are responsible for their expression. The site of a particular genetic system on a chromosome is called a locus. Each locus may be the site of several alleles (alternative genes). In an ordinary cell of the human body there are 46 chromosomes arranged in 23 pairs, 22 pairs of which are autosomes (chromosomes other than sex chromosomes), with the remaining pair being the sex chromosomes, designated XX in females and XY in males. The loci of the blood group systems are on the autosomes, except for Xg, which is unique among the blood groups in being located on the X chromosome. Genes carried by the X chromosome are said to be sex-linked. Since the blood groups are inherited in a regular fashion, they can be used as genetic markers in family studies to investigate whether any two particular loci are sited on the same chromosome—*i.e.,* are linked. The genes sited at loci on the same chromosome travel together from parent to child, and, if the loci are close together, the genes will rarely be separated.

Loci that are farther apart can be separated by recombination. This happens when material is exchanged between homologous chromosomes (pair of chromosomes) by crossing over during the process of cell division. The reproductive cells contain half the number of chromosomes of the rest of the body; ova carry an X chromosome and spermatozoa an X or a Y. The characteristic number of 46 chromosomes is restored at fertilization. In a classical pedigree linkage study, all the members of a family are examined for a test character and for evidence of the nonindependent segregation of pairs of characters. The results must be assessed statistically to determine linkage. Individual chromosomes are identified by the banding patterns revealed by different staining techniques. Segments of chromosomes or chromosomes that are aberrant in number and morphology may be precisely identified. Other methods for localizing markers on chromosomes include somatic cell hybridization (cell culture with alignment of single strands of RNA and DNA) and use of DNA probes (strands of radiolabeled DNA). These methods are useful in classical linkage studies to locate blood group loci. The loci for many red cell groups have been found on chromosomes and in many cases have been further localized on a particular chromosome. Their chromosome assignments and linkage to genes as well as associated abnormalities are outlined in Table 4.

Table 4: Chromosome Location, Gene Linkage, and Disease Associated with Genes for Blood Group Systems

chromosome	gene	known linkage	associated disease
1	Rh	EL_1	elliptocytosis
	Rh	PGD	
	Rh	PGM	
	Rh	ENO	
	Rh	UMPK	
	Rh	FUCA	
1	Radin	Rh	
1	Scianna	Rh	
1	Fy	AMY_1	
	Fy	AMY_2	
4	MNSs	Gc	sclerotylosis
4	Gc	Dombrock	
6	Chido	HLA	
6	Rogers	HLA	
9	ABO	AK_1	nail-patella syndrome
19	Lewis	C_3	
19	Lutheran	secretor	
	Lutheran	APOE	
	Lutheran	PEPD	
	Lutheran	DM	myotonic dystrophy
	Lutheran	FHC	familial hypercholesterolemia
	Lutheran	NF	neurofibromatosis
19	H		
X	Xg		hemophilia

Genetic determinations

Genetic pathways. In some of the blood group systems the amount of antigen produced depends on the genetic constitution. The ABO blood group gene codes for a specific carbohydrate transferase enzyme that catalyzes the addition of specific sugars onto a precursor substance. As a new sugar is added, a new antigen is produced. Antigens in the MNSs blood system are the products of genes that control terminal amino acid sequence. The amount of antigen present may depend on the amount of gene product inherited or on the activity of the gene product (*i.e.,* transferase). The red cells of a person whose genotype is *MM* show more M antigen than do *MN* red cells. In the case of ABO, the same mechanism may also play a role in antigen expression, but specific activity of the inherited transferase may be more important.

The amount of antigen produced can also be influenced by the position of the genes. Such effects within a genetic complex can be due to determinants on the same chromosome—they are then said to be in cis—or to determinants on the opposite chromosome of a chromosome pair—in trans.

In the Rh combination cdE/cde, more E antigen is produced than in the combination cDE/cde. This may be due to the suppressor effect of D on E. An example of suppression in the trans situation is that more C antigen is detectable on the red cells from CDe/cde donors than on those of CDe/cDE people. The inheritance of the Rh system probably depends on the existence of operator genes, which turn the activity of closely linked structural genes on or off. The operator genes are themselves controlled by regulator genes. The operator genes are responsible for the quantity of Rh antigens, while the structural genes are responsible for their qualitative characteristics.

The detection of recombination (exchange of material between chromosomes) or mutation in human families is complicated by questions of paternity. In spite of the large number of families that have been studied, it is an extremely rare occurrence. The paucity of examples may indicate that the recombinant and mutation rate for blood group genes is lower than that estimated for other human genes.

Dizygotic and monozygotic twins

Twins and chimeras. Twins may be either dizygotic or monozygotic. Dizygotic twins result from two distinct fertilizations and are therefore as different as two ordinary siblings. Monozygotic twins result from fertilization of one egg by one sperm followed by cleavage of the zygote during early mitotic division. These siblings are genetically identical. It is not possible to prove that they are identical but, just as in paternity studies, a high statistical probability can be ascertained by using red cell groups and other antigen inheritance patterns. Some unexpected discoveries also have been described as the result of blood group analysis. In occasional pairs of dizygotic twins a dual population of red cells with different markers can be found. The phenomenon, first described in cattle, is believed to occur because of vascular anastamosis between twins (exchange of cells and graft tolerance).

Blood groups and anthropology. The human species is

Populations

composed of natural reproductive groups referred to as populations. The blood groups are found in all human populations but vary in frequency. An analysis of populations yields striking differences in the frequency of some blood group genes. Table 5 indicates only the major and characteristic differences. The presence of the Di^a antigen indicates a Mongolian origin; the frequency of the antigen is about 10 percent among the Japanese and Chinese, but it has not been found in Eskimos. The Fy(a−b−) phenotype is characteristic of blacks; Negro origin can also be recognized by the presence of the Js^a antigen, which forms part of the Kell system and the V antigen of the Rh system.

Table 5: Distribution of Blood Groups in Populations

system	blood group	comments on frequency in populations
ABO	A	increased in Scandinavians, Australians, and North American Indians
ABO	B	increased in Central Asians; lowest frequency in Basques
ABO	O	increased in Central and South American Indians and northwest Europeans
ABO	A$_2$	present only in Europeans, Middle Easterners, and East Africans
ABO	Hh	silent varieties, (h) present in people of India
P	P$_1$	increased in black Africans; decreased in Far Easterners
Duffy	Fy(a−b−)	increased in black Africans and blacks
Kell	Js^a	increased in black Africans and blacks
Rh	V	increased in black Africans and blacks
Kidd	JK(a−b−)	increased in Filipinos and Hawaiians
Diego	Di^a	increased in Mongolians
MNSs	M, N, S, s	some rare antigens found only in blacks (S^u and Henshaw genes)

When a population is being investigated, an adequate sample of the people is classified according to the phenotypes. The gene and genotype frequencies can be computed according to the Hardy–Wienberg law, a mathematical model based on the assumptions that genes remain stable, matings take place randomly, and sample size is large.

Distribution of ABO groups

The frequency of the *A* gene is high in western Europe, western Asia, and among Australian Aborigines and certain American Indian tribes in North America. The *O* gene frequency is high in northern and western Europe and southern West Africa. Isolated populations such as the Indians of South and Central America may be entirely of group O. The maximum frequency of the *B* gene occurs in Central Asia and northern India. The *B* gene was probably absent from American Indians and Australian Aborigines before racial admixture occurred with the coming of the white man.

On the Rh system most northern and central European populations differ from each other only slightly and are characterized by a *cde* (*r*) frequency of about 40 percent. The frequencies of the most common Rh combinations of the English population are given in Table 6; they are representative of all Caucasian populations. The peoples of Africa show a preponderance of the complex *cDe* (R^0), and the frequency of *cde* (*r*) is about 20 percent. In eastern Asia, *cde* (*r*) is almost wholly absent, and, since everyone has the D (Rh$_o$) antigen, hemolytic disease of the newborn, due to the presence of maternal anti-D, is unknown in these populations.

Table 6: Frequency of the Most Common Rh Gene Complexes in the English Population

gene complex	frequency‡
*CDe** (R^1)†	41
cde (*r*)	39
cDE (R^2)	14
cDe (R^0)	3
C^wDe (R^{1w})	1
cdE (*r″*)	1
Cde (*r′*)	1
CDE (R^z)	<1

*Fisher–Race notation. †Wiener notation.
‡In percentage.

The uniqueness of the genetic characters of Africans probably results from their long isolation from other humans. The blood group frequencies in small inbred populations reflect the influences of genetic drift. In a small community an allele can be lost from the genetic pool if persons carrying it happen to be infertile, while it can increase in frequency if advantage exists. It has been suggested, for example, that *B* alleles were lost by chance from American Indians and Australian Aborigines when these communities were small. There are pronounced discrepancies in blood group frequencies between the people of eastern Asia and the aboriginal people of America. Other blood group frequencies in different populations show that ancestors might share some common attribute indicating a close resemblance between populations.

The ABO system and primates

Nonhuman primates carry blood group antigens that can be detected with reagents used for typing human beings. The closer their evolutionary relationship to humans, the greater their similarity with respect to antigens. The red cells of the apes, with the exception of the gorilla, have ABO antigens that are indistinguishable from those of human cells. Chimpanzees and orangutans are most frequently group A, but groups O, B, and AB are represented. Gibbons can be of any group except O, and gorillas have a B-like antigen that is not identical in activity with the human one. In both Old and New World monkeys the red cells do not react with anti-A or with anti-B, but, when the secretions are examined, A and B substances and agglutinins are present in the serum. The red cells of the apes have been shown to carry M-like and N-like antigens.

As far as the Rh system is concerned, chimpanzees carry two Rh antigens—D (Rh$_o$) and c (hr′)—but not the others, whereas gibbons have only c (hr′). The red cells of monkeys do not give clear-cut reactions with human anti-Rh sera. The original Rh antibody, anti-D (anti-Rh$_o$), was formed when rabbits and guinea pigs were immunized with the cells of rhesus monkeys, and the specificity appeared to be the same as the antibody formed in patients, which was also called anti-D (anti-Rh$_o$). Later, it became apparent that the antibodies from the two different sources were not developing in response to the same antigen.

By the time this was discovered it was too late to change the name Rh, important in clinical medicine. It was therefore decided that the antigen defined by the nonhuman antibody should be called LW, thus honouring Landsteiner and Wiener, and that Rh would define the blood group. The gene locus for LW is separate from that for Rh, but the two systems are in competition for a precursor substance. Most people have the LW antigen, but a very few do not.

Blood groups and biology. The blood groups are an important means of identifying an individual, and they play a role in linkage studies. Their relevance in medicine lies in the practice of blood transfusion and in the effects of incompatibility between mother and child and the pathogenesis of hemolytic disease of the newborn. The fundamental role of blood groups is related to the more recent elucidation of their presence on specific parts of the red cell membrane. Deletions of membrane proteins (blood groups) have been shown to cause shortened survival of the red cell. Other receptor functions are evident from the studies of chronic granulomatous disease (inherited disorder of a leukocyte function) and invasion of red cells by malaria parasites. The importance of the genetic polymorphisms involved in blood grouping have not been made clear; however, the biologic function of certain blood groups is evident from the diseases that result in persons who lack the red cell membrane protein. (S.La./E.M.B.)

Bleeding and blood clotting

SIGNIFICANCE OF HEMOSTASIS

Components of hemostasis

The evolution of high-pressure blood circulation in vertebrates has brought with it the risk of bleeding after injury to tissues. Mechanisms to prevent bleeding (*i.e.*, hemostatic mechanisms) are essential to maintain the closed blood-circulatory system. Normal hemostasis is the responsibility of a complex system of three individual components: blood cells (platelets), cells that line the blood vessels (endothelial cells), and blood proteins (blood-clotting proteins). The blood platelet is a nonnucleated cell that circulates in the blood in an inactive, resting form. Endothelial cells line the wall of the blood vessel and inhibit blood from clotting on the vessel wall under normal conditions. Blood-clotting proteins circulate in the blood plasma in an inactive form, poised to participate in blood coagulation upon tissue injury. Blood-clotting proteins generate thrombin, an enzyme that converts fibrinogen to fibrin, and a reaction that leads to the formation of a fibrin clot.

The hemostatic mechanism involves three physiologically important reactions: (1) the formation of a blood clot, (2) the formation of a platelet plug, and (3) changes associated with the wall of the blood vessel after injury of its cells. In humans, defects in any of these processes may result in persistent bleeding from slight injuries, or, alternatively, in an overreaction that causes the inappropriate formation of blood clots (thrombosis) in blood vessels. When a blood vessel is injured, blood escapes for as long as the vessel remains open and the pressure within the vessel exceeds that outside. Blood flow can be stopped or diminished by closing the leak or by equalizing the pressure. The leak may be closed by contraction of the blood vessel wall or by the formation of a solid plug. Pressure may be equalized by an increase in external pressure as blood becomes trapped in the tissues (hematoma) or by a decrease in the intravascular pressure (the pressure within the blood vessel) caused by constriction of a supply vessel. The timing and relative importance of these events can vary with the scale of the injury. Bleeding from the smallest vessels can be stopped by platelet plugs; when bleeding is from larger vessels, blood clot formation is required; in still larger vessels the severe drop in pressure associated with shock is the last line of defense.

THE HEMOSTATIC PROCESS

Blood vessels that constitute the circulatory system include arterioles (the smallest arteries) and venules (the smallest veins) connected by capillaries (the smallest of all blood vessels). Blood cells, including red cells and platelets, normally have no tendency to adhere to each other or to the lining (endothelium) of the vessels. An injury too slight to rupture a vessel, however, may still bring about a hemostatic reaction that causes blood cells to adhere to each other. After minor tissue injury there may be partial vessel contraction and platelet adhesion in successive layers at the point of injury. A platelet mass is formed that grows until it blocks, or almost blocks, the vessel. Sometimes this platelet mass breaks down and then reforms, a cycle that repeats perhaps many times. These masses consist of minimally altered platelets. Even these slight injuries cause shedding of some endothelial cells from the vessel and the exposure of deeper layers to which the platelets adhere.

Slight injuries

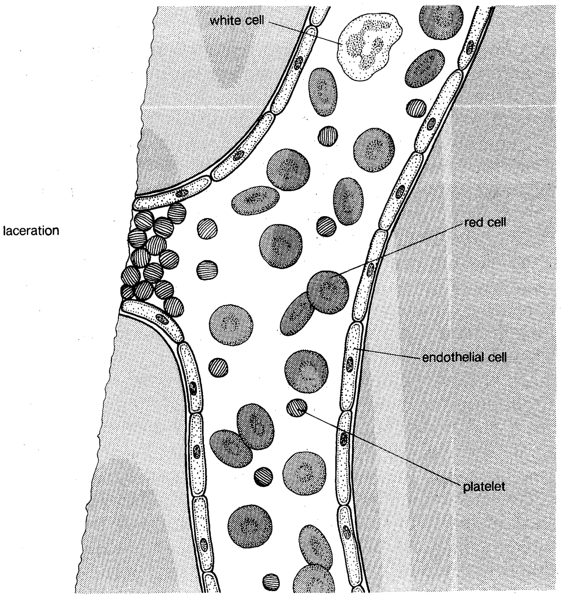

Figure 7: *Formation of the platelet plug.*
With laceration of a small blood vessel, platelets are recruited to the site of tissue injury. The clump of platelets is known as the platelet plug; the endothelial cells line the wall of the vessel.

Massive injuries

If the vessel is cut so that blood escapes, the hemostatic reaction is different. In muscular vessels there may be immediate contraction and narrowing of the vessel, but this usually only minimizes blood loss. A mass of activated platelets adheres to the site of vessel injury (a platelet plug) and normally stops the flow of blood out of the vessel (Figure 7). Unlike the platelets circulating in the blood and those adhering to minor tissue injuries, these platelets have undergone a biochemical and morphological change characteristic of platelet activation, a process that includes the secretion of the contents of platelet granules into the surrounding blood and the extension of pseudopodia. Between the platelets develop bundles of fibrin fibres (coagulation). These changes occur near damaged collagen, the fibrous protein found in connective tissue that underlies the endothelial cell. Later, normal healing of the wound occurs. The platelets subsequently degenerate into an amorphous mass and after several days, the fibrin itself is dissolved (fibrinolysis) by an enzyme, plasmin. The fibrin clot is replaced by a permanent framework of scar tissue that includes collagen, and healing is thus complete.

The normal hemostatic response to damage to the vascular endothelium can be organized into four stages: (1) initial vasoconstriction, (2) aggregation of platelets on and around the lesion and the formation of a platelet plug, (3) activation of the reactions of coagulation, and (4) the activation of fibrinolysis.

Vascular function. The most obvious hemostatic vascular reaction is constriction of the blood vessel after injury. This is important in large arteries because platelet adhesion and clotting are insufficient to arrest bleeding. Delayed surgical aid notwithstanding, the survival of some persons who have lost limbs in accidents is due to constriction of their main arteries. Other vascular reactions to injury have only a subsidiary hemostatic effect.

Platelets and their aggregation. Mammalian platelets are nonnucleate cells produced by large bone marrow cells called megakaryocytes and circulate in the blood in a resting, inactive form for an average of 10 days. The normal platelet count in humans is between 150,000 and 400,000 platelets per cubic millimetre of blood. The inactive platelet contains three types of internal granules: the alpha granules, the dense granules, and the lysosomes. Each of these granules is rich in certain chemicals that have an important role in platelet function. For example, dense granules contain large quantities of calcium ions and adenosine diphosphate (ADP). Upon release from the platelet, ADP stimulates other platelets to activate when it binds to the ADP receptor on the platelet membrane. The alpha granules contain many proteins, including fibrinogen, thrombospondin, fibronectin, and von Willebrand factor. Upon platelet activation, platelets alter their shape from discoid to spherical and extend long footlike projections called pseudopodia. The alpha granules and dense granules move to the surface of the platelet, fuse with the platelet membrane, and release their contents into the blood surrounding the platelet. The lysosomes contain enzymes that digest spent proteins and other metabolites of the cell.

Platelet granules

Activated platelets strongly adhere to surfaces other than the lining of blood vessels, such as collagen, glass, metals, and fabrics. Adherent platelets themselves become adhesive for other activated platelets so that, in a flow system, a platelet plug develops. The propagation of this adhesiveness from one layer to the next is probably due to chemicals, such as ADP and thromboxane A_2, secreted into the blood from the granules of the activated platelets. The ADP released from the dense granules binds to a receptor on the platelet surface, initiating the biochemical and morphological changes associated with platelet activation and secretion. The property of adhesiveness for normal platelets requires a protein on the surface of the platelet membrane, known as glycoprotein Ib, to bind von Willebrand factor, a large multimeric plasma protein released from the alpha granules. Von Willebrand factor,

The platelet plug

when bound to glycoprotein Ib on the platelet surface, facilitates the interaction of platelets with a variety of other surfaces (*e.g.,* the damaged vessel lining).

Platelet aggregation is the property of platelets to clump with each other to form a platelet plug. Two proteins on the platelet membrane play an important role in platelet aggregation: glycoprotein IIb and glycoprotein IIIa. These proteins form a complex in the membrane and expose a receptor site after platelet activation that binds fibrinogen (a bivalent molecule with two symmetrical halves that is found in relatively high concentration in plasma). Fibrinogen can bind simultaneously to two platelets. Thus, fibrinogen links platelets together (aggregation) through the glycoprotein IIb–IIIa complex that serves as the fibrinogen receptor.

Injury to the vessel lining and contact of the blood with tissues outside the vessel stimulates thrombin production by the activation of the clotting system. Thrombin causes platelet aggregation. Platelets exposed to thrombin secrete their granules and release the contents of these granules into the surrounding plasma.

Blood coagulation. Coagulation is the replacement of a relatively unstable platelet plug with a stronger, more resilient blood clot through a series of interdependent, enzyme-mediated reactions that bring about the generation of thrombin and the formation of fibrin from fibrinogen. The intrinsic and the extrinsic pathways of coagulation are involved in regulating coagulation; each is activated by a different trigger, although they share many steps in the course of the generation of thrombin.

Elements of the intrinsic pathway

Intrinsic pathway of blood coagulation. All the components necessary for the clotting process to proceed are found in the blood. As such, the proteins required for such clotting to take place are part of the intrinsic pathway of blood coagulation. This pathway involves a series of proteins, protein cofactors, and enzymes, which interact in reactions that take place on membrane surfaces. These reactions are initiated by tissue injury and result in the formation of a fibrin clot (Figure 8).

The intrinsic pathway is initiated by the activation of factor XII by certain negatively charged surfaces, including glass. High-molecular-weight kininogen and prekallikrein are two proteins that facilitate this activation. The enzyme form of factor XII (factor XIIa) catalyzes the conversion of factor XI to its enzyme form (factor XIa). Factor XIa catalyzes the conversion of factor IX to the activated form, factor IXa, in a reaction that requires calcium ions. Factor IXa assembles on the surface of membranes in complex with factor VIII; the factor IXa–factor VIII complex requires calcium to stabilize certain structures on these proteins associated with their membrane-binding properties. Factor X binds to the factor IXa–factor VIII complex and is activated to factor Xa. Factor Xa forms a complex with factor V on membrane surfaces in a reaction that also requires calcium ions. Prothrombin binds to the factor Xa–factor V complex and is converted to thrombin, a potent enzyme that cleaves fibrinogen to fibrin, a monomer. The monomer fibrin molecules then link together (polymerize) to form long fibres. Later, additional bonding between the units of the polymer is promoted by an enzyme known as factor XIIIa, which stabilizes the newly formed clot by cross-linkages. Although the detailed mechanisms are not known, this cascade, or waterfall, effect offers the possibility for amplification of a small signal associated with tissue injury into a major biologic event—the formation of a fibrin clot. Furthermore, careful regulation of this system is possible with the participation of two protein cofactors, factor VIII and factor V.

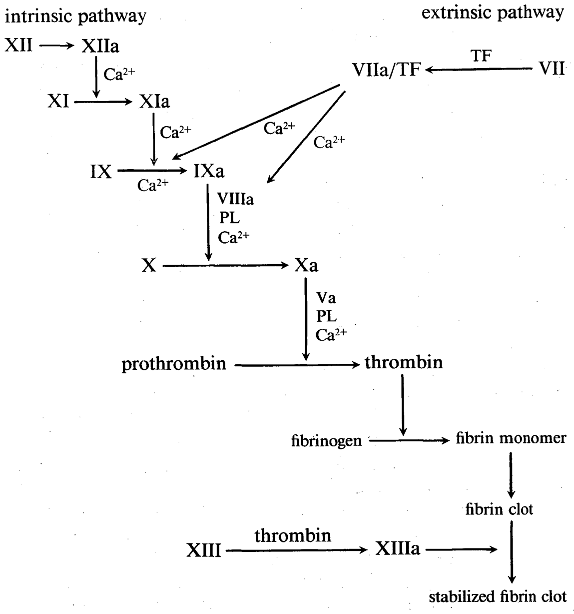

Figure 8: *The blood coagulation cascade.*
Each protein circulates in the blood in an active form. Upon initiation of the clotting system, the pathways for generating fibrin are activated. Many of the clotting proteins are referred to as numbered factors. PL indicates platelet phospholipids and TF indicates tissue factor.

Active and inactive surfaces

Certain negatively charged surfaces, including glass, kaolin, some synthetic plastics, and fabrics, activate factor XII to its enzyme form, factor XIIa. In contrast, certain materials have little tendency to activate factor XII. Inactive surfaces include some oils, waxes, resins, silicones, a few plastics, and endothelial cells, the most inert surface of all. The physicochemical properties that determine activity are not known. The problem is important, for modern surgery requires a perfectly inactive material to make substitutes (prostheses) for heart valves and sections of blood vessels. The formation of clots (thrombi) on these surfaces can lead to serious or even fatal complications. Open-heart surgery requires pumping of blood through equipment that does not activate the blood-clotting process significantly. Similarly, blood filtration of waste products during kidney dialysis must not lead to the generation of fibrin clots. To minimize the activation of blood coagulation when blood flows over foreign surfaces, special drugs (anticoagulants) such as heparin are employed.

The activity of the intrinsic pathway may be assessed in a simple laboratory test called the partial thromboplastin time (PTT), or, more accurately, the activated partial thromboplastin time. Plasma is collected and anticoagulated with citrate buffer; the citrate binds and effectively removes functional calcium ions from the plasma. Under these conditions, a fibrin clot cannot be generated. A negatively charged material, such as the diatomaceous material kaolin, is added to the plasma. Kaolin activates factor XII to its enzyme form, factor XIIa, which then activates factor XI. The process is blocked from further activation because of the lack of calcium ions, which are required for the next reaction, the activation of factor IX. Upon the addition of calcium ions and a phospholipid preparation (which serves as an artificial membrane for the assembly of the blood-clotting protein complexes), the duration of time is recorded until a visible clot is formed. This reaction takes place in a matter of 25 to 50 seconds, depending upon the formulation of chemicals used. In practice, the clotting time of a test plasma is compared to the clotting time of normal plasma. Delayed clotting, measured as a prolonged partial thromboplastin time, may be due to a deficiency in the activity of one or more of the blood-clotting factors or to a chemical inhibitor of blood coagulation.

Tissue factor

The extrinsic pathway of blood coagulation. Upon the introduction of cells, particularly crushed or injured tissue, blood coagulation is activated and a fibrin clot is rapidly formed. The protein on the surface of cells that is responsible for the initiation of blood clotting is known as tissue factor, or tissue thromboplastin. Tissue factor is found in many of the cells of the body but is particularly abundant in those of the brain, lungs, and placenta. The pathway of blood coagulation activated by tissue factor, a protein extrinsic to blood, is known as the extrinsic pathway (Figure 8).

Tissue factor serves as a cofactor with factor VII to facilitate the activation of factor X. Alternatively, factor VII can activate factor IX, which, in turn, can activate factor X. Once activated, factor X proceeds to activate prothrombin to thrombin in a reaction requiring factor V. The thrombin converts fibrinogen to fibrin. With the exception of factor VII, all components of the extrinsic pathway are also components of the intrinsic pathway.

The activity of the extrinsic pathway may be assessed in the laboratory using a simple test known as the prothrombin time. Tissue extract, or tissue thromboplastin, is extracted from animal tissues rich in tissue factor. Plasma, anticoagulated with citrate buffer, is allowed to clot with the simultaneous addition of phospholipid, calcium, and thromboplastin. The duration of time until clot formation, known as the prothrombin time, is usually between 10 and 12 seconds. In practice, the clotting time of a test plasma is compared to the clotting time of normal plasma. Delayed clotting, measured as a prolonged prothrombin time, may be due to a deficiency in the activity of one or more of the blood-clotting factors in the extrinsic pathway or to a chemical inhibitor of blood coagulation that interferes with the extrinsic pathway.

In summary, there are two independent mechanisms for initiating blood coagulation and for activating factor X: (1) negatively charged surfaces that initiate blood clotting through the intrinsic pathway (factors XII, XI, IX, and VIII), and (2) tissue factor on cells outside the blood that participates in the extrinsic pathway (factor VII). The common pathway (factor X, factor V, prothrombin, and fibrinogen) is shared by both systems. Although both pathways provide the opportunity to acquire meaningful information about clotting proteins using the partial thromboplastin time and the prothrombin time, it is most likely that the physiologically important pathway of blood coagulation is the extrinsic pathway initiated by tissue factor.

Zymogens

Biochemical basis of activation. The blood-clotting proteins circulate in the blood in their inactive, proenzyme form. The biochemical term for such proenzymes is zymogen. These zymogens are precursor enzymes that are converted to active enzymes by the cleavage of one or in some instances two peptide bonds. By splitting the protein into specific fragments, the zymogen is turned into an active enzyme that can itself split particular peptide bonds. This process, known generally as limited proteolysis, is equivalent to a molecular switch; by cutting a specific bond that connects two amino acids in the string of amino acids known as a polypeptide, an active enzyme is formed. Thus, the blood contains a system poised to become engaged instantaneously in the formation of blood clots if tissue is injured. Under normal conditions, however, blood clotting does not take place in the absence of tissue injury. The clotting proteins that function as zymogens in the blood include factor XII, factor XI, prekallikrein, factor IX, factor X, factor VII, and prothrombin.

Protein cofactors also play an important role in blood coagulation. Two protein cofactors, factor V and factor VIII, are large proteins that probably regulate blood coagulation. These proteins circulate in the blood as inactive cofactors. By the process of limited proteolysis, in which several cuts in the polypeptide chains of these cofactors are formed by the enzyme thrombin, factors V and VIII are converted to active cofactors. Factor V and factor VIII bind to membrane surfaces and form a focal point for the organization of certain protein complexes.

Antithrombin III

Inhibition of clotting. After the activation of the blood-clotting system, the active enzymes must be turned off and the clotting process contained locally to the area of tissue injury. The details of the regulation of blood coagulation remain obscure, but it is clear that a series of blood proteins play a specialized role in disengaging the activated blood-clotting system. Antithrombin III is a plasma protein that combines with thrombin as well as most of the other activated blood-clotting proteins (*e.g.,* factors Xa and IXa) to form inert complexes. This action is greatly enhanced by the presence of heparin, a substance formed by mast cells of the connective tissue. The hereditary deficiency of antithrombin III is associated with an excessive tendency toward clot formation, and manifestations of this defect are recurrent thrombophlebitis and pulmonary embolism. Heparin cofactor II is another plasma protease inhibitor that specifically forms a complex with thrombin, thus inactivating this enzyme. Protein C, a vitamin K-dependent protein, is a zymogen that requires vitamin K for its activation by thrombin complexed to thrombomodulin, a protein on the endothelial cell membrane. Activated protein C is capable of inactivating the active cofactor forms of factors VIII and V. Its action is enhanced when bound to protein S, a vitamin K-dependent protein that is attached to cell membranes (platelet or possibly endothelial cells). A deficiency in the level of protein C or protein S is associated with an excessive tendency to form clots.

Fibrinolysis

Another anticoagulant effect is the fibrinolytic (fibrin-splitting) action of plasmin, an enzyme that catalyzes the removal of old fibrin at injury sites and any which may be deposited in normal vessels. Plasmin is derived from plasminogen, an inert protein precursor that can be activated by tissue plasminogen activator. Streptokinase, urokinase, and tissue plasminogen activator are drugs that activate plasminogen and lead to the dissolution of clots.

Synthesis of blood-clotting proteins. Most of the blood coagulation proteins are synthesized in the liver. In addition, factor VIII is synthesized in a large number of other tissues. Six proteins involved in blood coagulation require vitamin K for their complete synthesis: factor IX, factor X, prothrombin, factor VII, protein C, and protein S. These proteins are synthesized in precursor form. In a region of the liver cell called the rough endoplasmic reticulum, specific glutamic acid residues in the protein are changed by an enzyme-mediated reaction to form a modified glutamic acid known as γ-carboxyglutamic acid. This enzyme reaction, known as γ-carboxylation, requires vitamin K as a cofactor. γ-Carboxyglutamic acid is a unique amino acid that binds to calcium. In the protein, γ-carboxyglutamic acids form the calcium-binding sites that characterize this form of calcium-binding protein, the vitamin K-dependent proteins. Calcium stabilizes certain structural forms of the vitamin K-dependent proteins, enabling these proteins to bind to cell membranes. In the absence of vitamin K or in the presence of vitamin K antagonists such as warfarin, γ-carboxylation is inhibited and proteins are synthesized that are deficient in γ-carboxyglutamic acid. These proteins have no biologic activity because they do not bind to calcium and do not interact with membrane surfaces. (B.Fu.)

Blood diseases

The hematopoietic system

The blood diseases are abnormal conditions that involve the corpuscular elements of the blood—red cells, white cells, and platelets—and the tissues in which they are formed—the bone marrow, the lymph nodes, the spleen (*i.e.,* the hematopoietic system). This definition, however, needs expansion and modification. In certain types of illness (*e.g.,* pneumonia, appendicitis) the number of leukocytes is increased. This is called leukocytosis and is a physiological response rather than a disease of the blood. In other diseases, as will be mentioned below, a reduction in the number of red cells in the blood (anemia) occurs; such anemia is not usually thought of as representing a "blood disease," but it does represent the response of the hematopoietic system to the underlying disease. Strictly speaking, the term blood disease refers only to those types of anemia and those disorders of leukocytes, of platelets, of coagulation, and of the bone marrow, lymph nodes, and spleen in which the blood-forming organs or coagulation systems are the primary sites involved.

In the following, all varieties of alteration in the three corpuscular elements of the blood will be discussed. For convenience, disorders chiefly affecting the red cells, the leukocytes, the platelets, and the process of blood coagulation will be considered in turn, but, as will become apparent, alterations in disease do not necessarily occur in only one of these groups of blood cells; in its reactions in disease the hematopoietic system may demonstrate multiple changes.

Since the blood circulates throughout the body and carries nutritive substances as well as waste products, examination of it can be important in detecting the presence of disease. Examination of the blood may be considered in two categories; namely, the analysis of the plasma and the study of the corpuscles. Examination of the plasma includes measurement of plasma proteins, blood sugar, salts (which, being in solution and in the ionic state, are referred to as electrolytes), lipids, enzymes, urea, and various hormones. Such measurements are useful in the identification of diseases that are not classified as blood diseases—*e.g.,* diabetes, kidney disease, and thyroid disease. Special studies of plasma or its components can be carried out to determine the status of blood clotting.

Historical background

Long before the nature and composition of blood were known, a variety of complaints were attributed to disordered blood. The red cells were not recognized until the 17th century, and it was another 100 years before one of the forms of the white cells, the lymphocyte, and the clotting of blood were described. In the 19th century other forms of leukocytes were discovered, and a number of diseases of the blood and blood-forming organs were distinguished.

In the 19th century and also in the first quarter of the 20th century much attention was given to descriptions of the morphological changes—the changes in form and structure—that take place in the blood during disease and to the signs and symptoms of the various blood diseases. In the years that followed, a more physiological approach began to develop, concerned with the mechanisms underlying the development of disease and with the ways in which abnormalities might be corrected. Following World War II, progress was greatly accelerated by the strong financial support that medical science received during a particularly productive age of medical research.

The study of a particular instance of disease involves inquiry into the circumstances of its development, the symptoms, and the course of the illness (the history). A thorough physical examination of the affected person and specific laboratory tests are essential.

In the case of the blood, certain features of the physical examination are especially important in diagnosis. These include noting the presence or absence of pallor or, the opposite, an excess of colour; jaundice, red tongue, enlargement of the heart or liver; the presence or absence of small purple spots or larger bruises in the skin; enlargement of lymph glands (nodes); enlargement of the spleen; and tenderness of the bones.

Laboratory studies particularly valuable in diagnosis include (1) determination of the number and characteristics of red cells in the blood; *i.e.,* the existence of anemia or polycythemia, (2) study of the white cells, their number and their proportions as to type, (3) enumeration of the blood platelets and a study of the blood-clotting process, and (4) in many instances a study of the bone marrow. It is sometimes necessary to remove a lymph node for microscopic examination, and X-ray examinations may be necessary for the detection of organ or lymph node enlargement or bone abnormalities. The more unusual cases may require further examinations—*e.g.,* special serological (serum-related) or biochemical procedures or various measurements using radioactive isotopes to outline an organ or quantitate blood volume.

DISORDERS AFFECTING RED CELLS

Normal values for red blood cells

The quantity of red blood cells in normal human beings varies with age and sex as well as with external conditions, primarily atmospheric pressure. At sea level an average normal man has 5,400,000 red blood cells per cubic millimetre of blood. These carry an average of 16 grams of hemoglobin per 100 millilitres of blood. If such blood is centrifuged so that the red cells are packed in a special tube known as the hematocrit, they are found, on the average, to occupy 47 percent of the volume of the blood. In the average woman the normal figures are lower than this (red cell count 4,800,000; hemoglobin 14 grams; volume of packed red cells 42 percent). In the newborn infant these values are higher but decrease in the course of the first several weeks of postnatal life to levels below those of the normal woman; thereafter, they rise gradually. The differences in male and female blood begin to appear at about the time of puberty. It should be emphasized that these figures represent average values; those found in normal persons range approximately 15 percent on either side of this mean.

From the physiological standpoint it is the quantity of hemoglobin in the blood that is important because this iron-containing protein is required for the transport of oxygen from the lungs to the tissues. In disease, as well as in certain situations in which physiological adjustments take place, the quantity of hemoglobin may be reduced below normal levels, a condition known as anemia, or may be increased above normal, leading to erythrocytosis (also called polycythemia).

The anemias. Anemia varies in severity, and the tolerance of different persons for anemia varies greatly, depending in part upon the rate at which it has developed. When anemia has developed gradually, affected persons may endure severe grades of anemia with few or no complaints, whereas rapidly developing anemia causes severe symptoms; if sufficiently severe and rapid in development, anemia can be fatal. In anemia the blood is capable of carrying only a reduced amount of oxygen to tissues, a condition that stimulates the lungs to increase the respiratory rate in order to pick up more oxygen and the heart to increase its rate (pulse) in order to increase the volume of blood delivered to the tissues. Associated complaints include a pounding headache due to the increased blood flow.

Determination of anemia

Anemia is always a sign, either predominant or incidental, of some underlying congenital condition or acquired disease. There are many varieties of anemia. Their clinical manifestations are generally similar, and yet they must be differentiated because their causes differ, and consequently their treatment is not the same. Differentiation is based on the history and physical examination, which may reveal an underlying cause, and on examination of the blood. The latter includes measurement of the degree of anemia and microscopic study of the blood cells. If the number of red cells, the hemoglobin concentration of the blood, and the volume of packed red cells are known, the mean volume and hemoglobin content can be calculated. The mean corpuscular volume (MCV) normally is 82 to 92 cubic micrometres, and about a third of this is hemoglobin (mean corpuscular hemoglobin concentration, MCHC, normally is 32 to 36 percent). If determined accurately, the MCV and the MCHC are useful indexes of the nature of an anemia.

Under the microscope the red cells of humans appear as round biconcave disks of uniform size with an average diameter of approximately 7.8 micrometres, or 0.0078 millimetre (a common pin is about one millimetre thick). Microscopic inspection of films of blood dried on glass slides and stained with aniline dyes allows observation of variations in the size and colour and other abnormalities of individual red cells and also permits examination of the white cells and platelets.

Normal production of red cells

Derangement of function in anemia. Red cells are formed within the marrow cavities of the central bones of the adult human skeleton (skull, spine, ribs, breastbone, pelvic bones). The marrow contains nucleated red cells (normoblasts), as well as white cells of all stages of development and megakaryocytes, the source of blood platelets. The normoblasts are present in various stages of development toward the mature, adult, nonnucleated, hemoglobin-containing red cells that will be released into the circulating blood.

The newly arrived red cells in the circulation remain as reticulocytes, young red cells with a characteristic threadlike network, for two or three days. Each day's output of new red cells survives an average of 120 days before succumbing to old age. The red cells are able to withstand the vicissitudes of the circulation because glucose absorbed from the plasma is metabolized within the cell to supply energy in the form of adenosine triphosphate (ATP) as well as to provide reducing systems that protect against the products of oxidation (oxidation-reduction systems).

Ultimately the aged red cells are broken down by specialized reticuloendothelial cells that are found throughout the body and especially in the spleen and liver. The hemoglobin is digested into its components: iron, a red pigment with a ring-shaped structural formula (porphyrin), and a protein (globin). The iron remains within the body to be used over and over again in the formation of new hemoglobin. The porphyrin ring opens and is changed chemically to become the yellow pigment of the blood plasma, bilirubin. This is then excreted by the liver and gives the bile its characteristic colour. The globin is metabolized. In a healthy person, red cell production (erythropoiesis) is so well adjusted to red cell destruction that the levels of red cells and hemoglobin remain constant.

The circulation is a closed system from which there normally is no loss of blood except that which occurs physiologically in menstruation. Anemia results when (1) the production of red cells and hemoglobin lags behind the normal rate of their destruction, (2) excessive destruction exceeds production, or (3) blood loss occurs. The bone marrow normally is capable of increasing production as much as sixfold to eightfold through an increased rate of development from the primitive precursors. Anemia ensues when the normal fine balance between production, destruction, and physiological loss is upset and erythropoiesis has not been accelerated to a degree sufficient to reestablish normal blood values.

Conditions of anemia

The rate of production of red cells by the bone marrow normally is controlled by a physiological feedback mechanism analogous to the thermostatic control of temperature in a room. The mechanism is triggered by a reduction of oxygen in the tissues (hypoxia) and operates through the action of a hormone, erythropoietin, in the formation of which the kidney plays an important role. Erythropoietin is released and stimulates further erythropoiesis. When oxygen needs are satisfied, erythropoietin production is reduced and red cell production diminishes.

Failure of production of red cells may be caused by deficiency of certain essential materials, such as iron, folic acid, or vitamin B_{12}. It may be due to other causes, such as the presence of certain types of disease—*e.g.,* infection; damage of the bone marrow by ionizing radiation or by drugs or other chemical agents; or anatomical alterations in the bone marrow, as by leukemia or tumour metastases (migration of tumour cells to the marrow from distant sites of origin). Accelerated destruction of red cells may occur for any one of a large variety of causes (see below *Hemolytic anemias*). Finally, blood loss may result from trauma or may be associated with a variety of diseases.

Persons whose anemia is due to increased destruction of red cells have excessive amounts of bilirubin in the plasma. They appear to be slightly jaundiced, and the excess pigments darken the excreta. Certain laboratory tests measure the degree of excessive pigment production. The bone marrow responds to increased destruction of red cells by increasing the rate of their production, thereby increasing the number of reticulocytes in the blood. These cells, in addition to their unique staining characteristics, are larger than fully mature red cells. If their number is increased sufficiently, the mean corpuscular volume of the cells in the circulation is increased. The anemia is then characterized as macrocytic.

Classification of anemia

Macrocytic anemia also is produced when the anemia results from impaired production of red cells, *e.g.,* when vitamin B_{12} or folic acid is lacking. In other circumstances, as for example when there is a deficiency of iron, the circulating red cells are smaller than normal and poorly filled with hemoglobin—this is termed hypochromic microcytic anemia. In still other forms of anemia there is no significant alteration in the size, shape, or coloration of the red cells—normocytic anemias.

Anemias may be classified according to the underlying abnormality in the basic physiological mechanism (decreased production, increased destruction, blood loss) or on morphological grounds (macrocytic, normocytic, or microcytic hypochromic) or according to their cause (*e.g.,* vitamin B_{12} deficiency). In practice it is by a combination of clinical, morphological, and physiological studies that the cause is determined. Accurate diagnosis is essential before treatment is attempted because, just as the causes differ widely, so the treatment of anemia differs from one patient to another. Indiscriminate treatment by the use of hematinics (drugs that stimulate production of red cells or hemoglobin) is wasteful and can be dangerous.

Megaloblastic anemias. Lack of vitamin B_{12} or folic acid leads to the production in the bone marrow of abnormal nucleated red cells known as megaloblasts. Such cells can be identified by their characteristic appearance. When such a vitamin deficiency occurs, bone marrow activity is seriously impaired; marrow cells proliferate but do not mature properly, and erythropoiesis becomes largely ineffective. Anemia develops, the number of reticulocytes is reduced, and even the numbers of granulocytes (leukocytes that contain granules in the cell substance outside the nucleus) and platelets are decreased. The adult red cells that are formed from megaloblasts are larger than normal, resulting in a macrocytic anemia. The impaired and ineffective erythropoiesis is associated with accelerated destruction of the red cells, thereby providing the features of a hemolytic anemia.

Vitamin B_{12} is a red, cobalt-containing vitamin that is found in animal but not in vegetable foods. Unlike other vitamins, it is not formed by higher plants but only by certain bacteria and molds and in the rumina of sheep and cattle, provided that traces of cobalt are present in their fodder. In other species, including humans, vitamin B_{12} must be obtained passively, by eating food of animal source. Furthermore, this vitamin is not absorbed efficiently from the human intestinal tract unless a certain secretion of the stomach, the so-called intrinsic factor (IF), is available to concentrate the vitamin on the intestinal wall.

Pernicious anemia

The most common cause of vitamin B_{12} deficiency is pernicious anemia, a condition usually affecting patients past middle age. In this disorder all stomach secretion of IF fails, perhaps as the result of an immune process consisting of the production of antibodies directed against the stomach lining. The tendency to form such antibodies may be hereditary.

The English physician Thomas Addison first described this condition in 1849. The discovery of vitamin B_{12} came about because of the investigations of the American physician George H. Whipple, who studied the value of various foods in promoting the formation of hemoglobin in dogs made anemic by bleeding. Whipple found liver to be most effective in the treatment of pernicious anemia. The American physicians George R. Minot and William P. Murphy further tested the value of liver. The activity of the foods tested by Whipple was due mainly to their iron content rather than to the presence of vitamin B_{12}. Investigations by William B. Castle, also an American physician and medical scientist, later revealed the defect in pernicious anemia to be a deficiency of a factor secreted by the stomach, which is necessary for the absorption of vitamin B_{12} across the wall of the intestine into the circulation.

The first treatment for pernicious anemia was to prescribe that the patient eat liver each day; later regular amounts of liver extract were given orally or by injection. Today patients are given injections of the equivalent of a millionth of a gram of vitamin B_{12} per day. In practice, the necessary amount of this vitamin can be given once a month or even once in three months. Oral treatment with vitamin B_{12} is possible but inefficient because absorption is poor.

Other forms of vitamin B_{12} deficiency are rare. They are seen in complete vegetarians whose diets lack vitamin B_{12}, in persons whose stomachs have been completely removed and so lack a source of IF, in those who are heavily infested with the fish tapeworm *Diphyllobothrium latum* or have intestinal cul-de-sacs or partial obstructions where competition by the tapeworms or by bacteria for vitamin B_{12} deprives the host, and in persons with primary intestinal diseases that affect the absorptive capacity of the small intestine (ileum). In these conditions, additional nutritional deficiencies, as of folic acid and iron, are also likely to develop.

Effects of folic acid deficiency

Blood changes similar to those occurring in vitamin B_{12} deficiency result from deficiency of folic acid. Folic acid is a vitamin found in leafy vegetables, but it is also synthesized by certain intestinal bacteria. In humans deficiency usually is the result of a highly defective diet or of chronic intestinal malabsorption as mentioned above. Pregnancy greatly increases the need for this vitamin. There is also an increased demand in cases of long-continued accelerated production of red cells. This type of deficiency also has been observed in some patients receiving anticonvulsants, and there is some evidence that absorption of the vitamin may be impaired in these cases. Often several factors affecting supply and demand of the vitamin play a role in producing folic acid deficiency.

Unless folic acid deficiency is complicated by the presence of intestinal or liver disease, its treatment rarely requires more than the institution of a normal diet. In any event the oral administration of folic acid relieves the megaloblastic anemia. Some effect can be demonstrated even in pernicious anemia, but this treatment is not safe because the nervous system is not protected against the effects of vitamin B_{12} deficiency, and serious damage to the nervous system may occur unless vitamin B_{12} is given.

In the above conditions, megaloblastic anemia develops as the result of dietary deficiency of, faulty absorption of, or increased demands for vitamin B_{12} and/or folic acid. In addition to these circumstances, megaloblastic anemia may arise in still other situations. Selective vitamin B_{12} malabsorption may be the consequence of a hereditary defect. Deranged metabolism may play a role in some instances of megaloblastic anemia that accompany pregnancy. Metabolic antagonism is thought to be the mechanism underlying the megaloblastic anemia associated with the therapeutic use of certain anticonvulsant drugs and some drugs employed in the treatment of leukemia and other forms of cancer. In fact, one of the earliest drugs used to treat leukemia was a folic acid antagonist.

Normocytic normochromic anemias. The term normocytic normochromic anemia is applied to those forms of anemia in which the mean size and hemoglobin content of the red cells are within normal limits. Usually microscopic examination of the red cells shows them to be much like normal cells. In other cases there may be marked variations in size and shape, but these are such as to equalize one another, thus resulting in normal mean values. The normocytic anemias are a miscellaneous group, by no means as homogeneous as the megaloblastic anemias.

Blood loss and chronic infections

Anemia caused by the sudden loss of blood is necessarily normocytic at first, since the cells that remain in the circulation are normal. The blood loss stimulates increased production, and the young cells that enter the blood in response are larger than those already present in the blood. If the young cells are present in sufficient number, the anemia temporarily becomes macrocytic (but not megaloblastic).

The many causes of acute blood loss include trauma, peptic ulceration of the stomach or bowel, and ulcerative lesions of other types. Treatment includes replacement, by transfusion, of the blood lost and appropriate steps directed against the cause.

A common form of anemia is that occurring in association with various chronic infections and in a variety of chronic systemic diseases. As a rule the anemia is not severe, although the anemia associated with chronic renal insufficiency (defective functioning of the kidneys) may be extremely so. Most normocytic anemias appear to be the result of impaired production of red cells, and in renal failure there is a deficiency of erythropoietin, the factor in the body that normally stimulates red cell production. In these states, the life span of the red cell in the circulation may be slightly shortened, but the mechanism of production of the anemia is failure of red cell production. The anemia of chronic disorders is characterized by abnormally low levels of iron in the plasma in the face of excessive quantities in the reticuloendothelial cells (cells whose function is ingestion and destruction of other cells and of foreign particles) of the bone marrow. Successful treatment depends on the possibility of eliminating or relieving the underlying disorder.

The mild anemias associated with deficiencies of the anterior pituitary, thyroid, adrenocortical, and testicular hormones usually are normocytic. As in the case of anemia associated with chronic infections or various systemic diseases, the symptoms usually are those of the underlying condition, although sometimes anemia may be the most prominent sign. Unless complicated by deficiencies of vitamin B_{12} or iron, these anemias are abolished by appropriate treatment with the lacking hormone.

Invasion of bone marrow by cancer cells carried by the bloodstream, if sufficiently great, is accompanied by anemia, usually normocytic in type but associated with abnormalities of both red and white cells. It is thought that such anemia is due to impaired production of red cells through mechanical interference. Whether this is true or not, a characteristic sign in the peripheral blood is the appearance of many irregularities in the size and shape of the red cells and of nucleated red cells; these young cells normally never leave the bone marrow but appear when the marrow's structure is distorted by invading cells.

In aplastic anemia the normally red marrow becomes fatty and yellow and fails to form enough of its three cellular products—red cells, white cells, and platelets. Anemia with few or no reticulocytes, reduced levels of the types of leukocytes formed in the bone marrow (granulocytes), and reduced platelets in the blood are characteristic. Manifestations of aplastic anemia are related to these deficiencies and include weakness, increased susceptibility to infections, and bleeding manifestations. In some cases the onset of aplastic anemia has been found to have been preceded by exposure to such organic chemicals as benzol, insecticides, or a variety of drugs, including especially the antibiotic chloramphenicol. While it is well established that certain agents may produce aplastic anemia, most persons exposed to these agents do not develop the disease and most persons with aplastic anemia have no clear history of exposure to such agents. There are other agents that produce aplasia in a predictable way. These include some of the chemotherapeutic agents used in the treatment of cancer, lymphoma, and leukemia, as well as irradiation treatment for these diseases. Because of this fact, blood counts are frequently checked and doses of drugs or irradiation modified in patients being treated. Withdrawal of medication is followed by recovery of the bone marrow in such cases. On the other hand, in those patients who develop aplastic anemia as a result of exposure to other toxic agents, cessation of exposure may not result in recovery of the marrow, or at best the marrow may be indolent and incomplete.

Chemotherapy and aplasia

Treatment of aplastic anemia is a twofold process. First, the patient must be supported through complications of the disease: infection calls for vigorous treatment with antibiotics; symptoms due to anemia call for red cell transfusions; bleeding calls for platelets. Second, efforts should be directed toward inducing bone marrow recovery. Based on the hypothesis that one of the mechanisms of production of the aplasia is autoimmunity, medication to suppress the immune response, such as the administration of antithymocyte globulin, is occasionally successful. An important and effective treatment is transplantation of bone marrow from a normal, compatible donor, usually a sibling. This treatment is limited by the availability of compatible donors and also by the fact that the recipient is increasingly prone to serious complications with advancing age.

Hypochromic microcytic anemias. Hypochromic microcytic anemias, characterized by the presence in the circulating blood of red cells that are smaller than normal and poorly filled with hemoglobin, fall into two main categories. The first is that due to deficiency of iron, and the second is that due to impaired production of hemoglobin (the globin); in either case there is an inadequate amount of the final product in the red cell.

Iron-deficiency anemia

Deficiency of iron is probably the most common cause of anemia throughout the world. Iron is required for hemoglobin formation; if the supply is insufficient to produce normal quantities of hemoglobin, the bone marrow ultimately is forced to produce cells that are smaller than normal and poorly filled with hemoglobin. Iron is derived from the diet and absorbed in the intestinal tract. Once in the body, it is retained and used over and over again, only minimal amounts being lost through shedding of cells from the skin and the exposed membranes and through normal menstruation. In the adult the body content is approximately 3.7 grams of iron, of which more than half is hemoglobin. In the male there is virtually no further need for iron. Deficiency results if the dietary supplies of iron are insufficient to meet the needs; if absorption is faulty, as in malabsorption disorders; or if blood loss is occurring. Common causes of iron deficiency are excessive menstrual loss in women and bleeding peptic ulcer in men. Iron deficiency is common in infancy and childhood because demands are great for the ever-expanding pool of circulating hemoglobin in the growing body and in pregnancy when the fetus must be supplied with iron. Hookworm infestation is a common cause of iron deficiency where conditions for the worm are favourable, because the intestinal blood loss caused by the myriad of worms attached to the wall is great.

Victims of iron-deficiency anemia are pale but not jaundiced. The deficiency of iron-containing enzymes in the tissues, if sufficiently great, results in a smooth tongue; brittle, flattened fingernails; and lustreless hair. Under the name of chlorosis, this type of anemia was mentioned in popular literature and depicted in paintings, especially those of the Dutch masters, until the 20th century. Although it is not necessarily less common now, there is no doubt that it is less severe in Europe and North America than it once was. The only treatment required is oral administration of iron salts in some palatable form, such as ferrous sulfate.

Small red cells poorly filled with hemoglobin are characteristic of a hereditary disorder of hemoglobin formation, thalassemia, that is common among people of Mediterranean stock and is discussed below. With the exception of iron deficiency and thalassemia, hypochromic microcytic anemia is rare. It is seen in anemia responsive to the vitamin pyridoxine, where the anemia probably results from a metabolic fault in the synthesis of the heme portion of hemoglobin. Sideroblastic anemia, characterized by the presence in the bone marrow of nucleated red cells, the nucleus of which is surrounded by a ring of iron granules ("ringed sideroblasts") and by a proportion of small, pale red cells in the blood, is of unknown cause and difficult to treat.

Sideroblastic anemia

Hemolytic anemias. Destruction of red cells at a rate substantially greater than normal, if not compensated for by accelerated red cell production, causes a type of anemia called hemolytic. Increased red cell destruction is recognized by demonstrating increased quantities of the pigmentary products of their destruction, such as bilirubin and urobilinogen, in the blood plasma, urine, and stools andbyevidence of accelerated erythropoiesis, such as an increase in the number of young corpuscles (reticulocytes) in the blood. When blood destruction is extremely rapid or occurs in the blood vessels, free hemoglobin is found in the urine (hemoglobinuria). Treatment varies with the cause of the hemolytic anemia.

There are two principal causes of hemolytic anemia: (1) inherently defective red cells and (2) an environment hostile to red cells. Abnormalities within the red cell are usually congenital and hereditary. They are exemplified by diseases in which the cell membrane is weakened, cell metabolism is defective, or hemoglobin is abnormal.

Hereditary defects

Hereditary spherocytosis is the most common disease involving the red cell membrane. It is characterized by the presence of red cells that appear small, stain densely for hemoglobin, and look nearly spherical. Such cells are mechanically fragile and readily swell up and burst in dilute salt solution. In the body they break up when deprived of free access to plasma glucose. The abnormality is aggravated by a tendency for the cells to remain longer than usual in the spleen because of their spheroidal shape. The corpuscular defect may appear if it is inherited from either parent (it is caused by a dominant gene). The anemia varies in severity. It may be so mild as to pass unnoticed for years, but it may suddenly become severe; *e.g.,* when an incidental respiratory infection briefly suppresses the accelerated production of red cells necessary to meet the constantly increased rate of their destruction. An agent known to cause this transient cessation of erythropoiesis is the parvovirus, and the development of severe anemia under these circumstances is termed aplastic crisis. Removal of the spleen, which always is enlarged, cures the anemia by eliminating the site of sequestration and destruction of the red cells but does not prevent hereditary transmission of the disease.

Red cells metabolize glucose by breaking it down to lactic acid either via an anaerobic (oxygenless) pathway, or by oxidation through a pathway called the pentose shunt. The anaerobic pathway, the main route of metabolism, provides energy in the form of ATP. Deficiencies of enzymes, such as pyruvate kinase, in this pathway shorten red cell survival times because energy-requiring activities within the red cell are curtailed. Deficiencies of enzymes in the anaerobic pathway, also a vital pathway for red cell survival, are generally relevant only when they are homozygous (*i.e.,* when the deficiency is inherited from each parent on an autosomal chromosome and is therefore expressed). Abnormalities also have been discovered in the alternative process of glucose metabolism, known as the phosphogluconate oxidative, or pentose shunt, glycolytic pathway. The pathway provides a reducing factor, NADPH (nicotinamide-adenine dinucleotide phosphate), which is part of a mechanism for returning glutathione to its reduced form following its oxidation by peroxides, a reaction that protects other molecules in the cell from oxidative damage by these peroxides. Deficiency of the first enzyme in the pathway, glucose-6-phosphate dehydrogenase (G6PD), is rather common. This deficiency results in a decreased capacity to detoxify peroxides and leads to oxidative damage to the red cell and hemolysis. If the deficiency is mild, there is hemolysis only when an excessive amount of peroxides is present, as may occur with the administration of some medications. The frequency of G6PD deficiency varies up to 36 percent among different Caucasian population groups and is especially frequent in Mediterranean peoples. Ingestion of fava beans by some of these people may induce severe destruction of red cells. The deficiency rarely is extreme and results in hemolysis without exposure to oxidants, medications, or toxins. Another variety of G6PD deficiency has been found in 10 to 14 percent of American blacks; the defect is harmless unless the person is exposed to certain drugs, such as certain antimalarial compounds (*e.g.,* primaquine) and sulfonamides. The full effect of the deficiency is rarely observed in females because the gene is sex-linked (*i.e.,* carried on the X chromosome), and only rarely do both X chromosomes carry the abnormal gene. Males, on the other hand, have only one X chromosome and thus only one gene available, and therefore the deficiency is fully expressed if it is inherited on the X chromosome from the mother.

Acquired forms of anemia

Hemolytic anemia can also result as the consequence of an environment hostile to the red cell. Certain chemical agents destroy red cells whenever sufficient amounts are given (*e.g.,* phenylhydrazine); others are harmful only to persons whose red cells are sensitive to the action of the agent. A number of the toxic drugs are oxidants or are transformed into oxidizing substances in the body. Injury may be accidental, as with moth ball (naphthalene) ingestion in children, or it may be the undesirable effect of a drug used therapeutically. Individual sensitivity is of several kinds. There is the susceptibility of certain patients to oxidant drugs such as antimalarial compounds mentioned above. This is attributable to a sex-linked, inherited deficiency of the enzyme G6PD. In other instances, sensitivity is on an immunologic basis; *e.g.,* hemolytic anemia caused by penicillin, quinidine, or alpha-methyldopa (a drug used in the treatment of elevated blood pressure). The anemia develops rapidly over a few days and without transfusions may be fatal.

A long recognized type of hemolytic anemia is that associated with the transfusion of incompatible red cells. Antibodies to the substances alpha- and beta-isoagglutinin, which occur naturally in the blood, destroy the donor red cells when incompatible blood is given by transfusion. Besides the best-known blood groups—A, B, and O—there are other groups to which a person may develop antibodies that will cause transfusion reactions. The rhesus (Rh) and Kell groups are examples. In hemolytic disease of the newborn (erythroblastosis fetalis) the destruction of the blood of the fetus by that of the mother may be due to Rh or ABO incompatibility. The events that take place are, first, the passage of incompatible red cells from the fetus into the circulation of the mother through a break in the placental blood vessels, then development of antibodies in the mother, and, finally, passage of these antibodies into the fetus, with consequent hemolysis, anemia, and jaundice.

Autoimmune hemolytic anemia

A form of hemolytic anemia that is relatively common depends on the formation of antibodies within the patient's body against his own red cells (autoimmune hemolytic anemia). This may occur in association with the presence of certain diseases, but it is often seen without other illness. Trapping of the red cells by the spleen is thought to depend on the fact that red cells coated with incomplete (nonhemolytic) antibody, when brought into contact with reticuloendothelial cells, adhere, become spherical, are ingested (phagocytosed), and break down.

Such anemias may be severe but often can be controlled by the administration of adrenocorticosteroids (which interfere with the destructive process) and treatment of the underlying disease, if one is present. In a number of instances, splenectomy—removal of the spleen—is necessary and is usually partially or wholly effective in relieving the anemia. The effectiveness of splenectomy is attributed to the removal of the organ in which red cells, coated with antibody, are selectively trapped and destroyed.

Other varieties of hemolytic anemia include that associated with mechanical trauma, such as that produced by the impact of red cells on artificial heart valves, excessive heat, and infectious agents (*e.g.,* the organism causing malaria).

Alpha and beta chains

Thalassemia and the hemoglobinopathies. Hemoglobin is composed of a porphyrin compound (heme) and globin. Normal adult hemoglobin (Hb A) consists of globin containing two pairs of chains of amino acids, of which the alpha chain consists of 141 amino acids, the beta chain 146. (A chain of amino acids is called a peptide or, alternatively, when many amino acids make up the chain, a polypeptide.) A minor fraction of normal adult hemoglobin consists of hemoglobin A_2, which contains alpha and delta chains. A different hemoglobin (Hb F) is present in fetal life and possesses a pair of the same alpha chains as does Hb A, but the second set is different (gamma chains). In normal hemoglobin the order in which the amino acids follow one another in the chain is always exactly the same. Studies by a number of investigators have now shown that qualitative and quantitative abnormalities in the globin chains can lead to disease.

In thalassemia it is thought that a primary genetic defect results in reduction in the rate at which alpha, beta, or delta chains are manufactured, the chains being otherwise normal. The relative deficiency of one pair of chains and the resultant imbalance of chain pairs results in ineffective production of red cells, deficient hemoglobin production, microcytosis (small cells), and hemolysis. In sickle-cell anemia and in other abnormalities affecting hemoglobin, the substitution of one amino acid for another at a particular site in the chain is the underlying fault. The substitution of valyl for glutamyl in the sixth position of the beta chain, for example, results in the formation of Hb S (the hemoglobin of sickle-cell disease) instead of Hb A. The defect is inherited as a Mendelian recessive. Thus, if only one parent transmits the abnormality, the offspring inherits the trait but is harmed relatively little; the red cells contain more Hb A than Hb S. If the trait is inherited from both parents, the predominant hemoglobin in the red cell is Hb S; the serious and sometimes fatal disease sickle-cell anemia is the consequence.

Since the first characterization of the nature of the abnormality in Hb S by Linus Pauling and his associates (1949), more than 100 abnormal hemoglobins have been identified, and other forms of "molecular disease" have been recognized as well. Fortunately, most abnormal hemoglobins are not sufficiently affected to alter their function, and therefore no observable illness occurs.

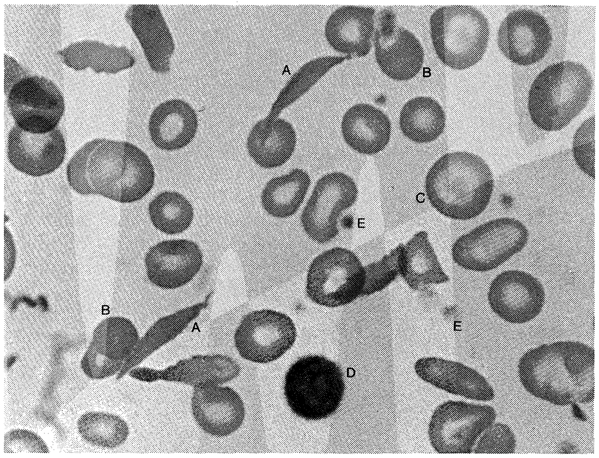

Figure 9: *Blood smear in which the red cells show variation in size and shape typical of sickle-cell anemia.*
(A) Long, thin, deeply stained cells with pointed ends are irreversibly sickled. (B) Small, round, dense cells are hyperchromic because a part of the membrane is lost during sickling. (C) Target cell with a concentration of hemoglobin on its centre. (D) Lymphocyte. (E) Platelets.

Sickle-cell trait and sickle-cell anemia

Sickle-cell anemia (Figure 9) occurs almost exclusively in blacks. In the United States at least 8 percent of blacks carry the sickle-cell trait. The actual disease, sickle-cell anemia, is less common (about 1 in 400 blacks). In this condition most of the red cells of a sample of fresh blood look normally shaped—discoidal—until deprived of oxygen, when the characteristic sickle- or crescent-shaped forms with threadlike extremities appear. Re-exposure to oxygen causes immediate reversion to the discoidal form. Sickle-cell anemia is characterized by severe chronic anemia, punctuated by painful crises, the latter being due to blockage of the capillary beds in various organs by masses of sickled red cells. This gives rise to fever and episodic pains in the chest, abdomen, or joints that are difficult to distinguish from the effects of other diseases. Death results from anemia, from infections, or, ultimately, from heart or kidney failure. While the many complications of the disease can be treated and pain relieved, there is no treatment to reverse or prevent the actual sickling process.

Thalassemia

Thalassemia (Greek: "sea blood") is so called because it was first discovered among peoples around the Mediterranean Sea, among whom its incidence is high. The thalassemias are another group of inherited disorders in which one or more of the hemoglobin subunits are synthesized defectively. This condition, when inherited from one parent, is called thalassemia minor; it causes serious disease only when inherited from both parents (thalassemia major, Cooley's anemia). Thalassemia now is known also to be common in Thailand and elsewhere in the Far East. The red cells in this condition are unusually flat with central staining areas and for this reason have been called target cells. In the mild form of the disease, thalassemia minor, there is usually only slight or no anemia, and life expectancy is normal. Thalassemia major is characterized by severe anemia, great enlargement of the spleen, and body deformities associated with expansion of the bone marrow. The latter presumably represents a response to the need for greatly accelerated red cell production by genetically defective red cell precursors, which are relatively ineffective in producing mature red cells. Anemia is so severe that transfusions are often necessary; however, they are of only temporary value and lead to excessive iron in the tissues once the transfused red cells break down. The enlarged spleen may further aggravate the anemia by pooling and trapping the circulating red cells. Splenectomy may partially relieve the anemia but does not cure the disease.

Causes of thalassemia

The defect in thalassemia may involve the beta chains of globin (beta-thalassemia), the alpha chains (alpha-thalassemia), the delta chains (delta-thalassemia), or both delta- and beta-chain synthesis. In the last (delta–beta-thalassemia), Hb F concentrations usually are considerably elevated since the number of beta chains available to combine with alpha chains is limited and gamma chain synthesis is not impaired. Beta-thalassemia comprises the majority of all thalassemias. A number of genetic mechanisms account for impaired production of beta chains, all of which result in inadequate supplies of messenger RNA available for proper synthesis of the beta polypeptide at the ribosome. In some cases no messenger RNA is produced. Most defects have to do with production and processing of the RNA from the beta gene; in alpha thalassemia, by contrast, the gene itself is deleted. There are normally two pairs of alpha genes, and the severity of the anemia is determined by the number deleted. Since all normal hemoglobins contain alpha chains, there is no increase in F or A$_1$. The extra non-alpha chains may combine into tetramers to form beta$_4$ (hemoglobin H) or gamma$_4$ (hemoglobin Barts). These tetramers are ineffective in delivering oxygen and are unstable. Inheritance of deficiency of a pair of genes from both parents results in intrauterine fetal death or severe disease of the newborn child (hydrops fetalis).

In most forms of hemoglobin abnormality only a single amino acid substitution occurs, but there may be combinations of hemoglobin abnormalities, or a hemoglobin abnormality may be inherited from one parent and thalassemia from the other. Thus, sickle-thalassemia and Hb E-thalassemia are relatively common.

Erythrocythemia

A malfunction of the abnormal hemoglobin may result in erythrocythemia, or overproduction of red cells. In these cases there is increased oxygen affinity, limiting proper delivery of oxygen to tissues and thereby stimulating the bone marrow to increase red cell production. In other cases the iron in heme may exist in the oxidized, or ferric (Fe^{3+}), state and thus cannot combine with oxygen to carry it to tissues. This results in a bluish colour of the skin and mucous membranes (cyanosis). The abnormality in the globin molecule that accounts for this is usually in an area of the molecule called the heme pocket, which normally protects the iron against oxidation, despite the fact that oxygen is being carried at this site.

Erythrocytosis. Erythrocytosis is an increase above normal in the number of red cells in the circulating blood, usually accompanied by an increase in the quantity of hemoglobin and in the volume of packed red cells. The increase may be either an actual rise in the total quantity of red blood cells in the circulation (absolute erythrocytosis), or it may be the result of a loss of blood plasma and thus a relative increase in the concentration of red cells in the circulating blood (relative erythrocytosis). The latter may be the consequence of abnormally lowered fluid intake or of marked loss of body fluid, such as occurs in persistent vomiting, severe diarrhea, or copious sweating or when water is caused to shift from the circulation into the tissue.

Absolute erythrocytosis occurs in response to some known stimulus for the production of red cells. This is in contrast to a disease called polycythemia vera, in which an increased amount of red cells are produced without a known cause. In polycythemia vera there is usually an increase in other blood elements as well.

Common causes

Erythrocytosis is a response by the body to an increased demand for oxygen. It occurs when hemoglobin is not able to pick up large amounts of oxygen from the lungs; *i.e.,* when it is not "saturated." This may result from decreased atmospheric pressure, as at high altitudes, or from impaired pulmonary ventilation. The sustained increase in red cells in persons who reside permanently at high altitudes is a direct result of the diminished oxygen pressure in the environment. Chronic pulmonary disease (*e.g.,* emphysema—abnormal distension of the lungs with air) may produce chronic hypoxemia (reduced oxygen tension in the blood) and lead to erythrocytosis. Extreme obesity also may severely impair pulmonary ventilation and thereby cause erythrocytosis (Pickwickian syndrome).

Congenital heart disorders that permit shunting of blood from its normal path through the pulmonary circuit, thereby preventing adequate aeration of the blood, can also cause erythrocytosis, as can a defect in the circulating hemoglobin. The latter defect may be congenital because of an enzymatic or a hemoglobin abnormality, as mentioned above; or it may be acquired as the result of the excessive use of coal-tar derivatives, such as phenacetin, which convert hemoglobin to pigments incapable of carrying oxygen (methemoglobin, sulfhemoglobin). Lastly, erythrocytosis can develop in the presence of certain types of tumours and as the result of the action of adrenocortical secretions. Treatment of erythrocytosis due to any of these causes involves the correction or alleviation of the primary abnormality.

Polycythemia vera

In polycythemia vera, the numbers of red cells, and often also the numbers of white cells and platelets, are increased and the spleen usually is enlarged. In this disease the stem cell precursor of the bone marrow cells produces excessive progeny. Afflicted persons have an exceptionally ruddy complexion and may complain of headaches, dizziness, a feeling of fullness, and other symptoms. Because of the excessive quantities of red cells, the blood is usually thick, and its flow is retarded; it sometimes clots in the blood vessels (thrombosis) of the heart, the brain, or the extremities with serious consequences. One of the simplest methods of treatment is to remove the blood, one pint at a time, from a vein until the cellular level approaches normal and the symptoms disappear. Occasionally it may be necessary to use drugs or radiation, in the form of radioactive phosphorus, to restrain the overactivity of the marrow cells. These treatments, however, must be avoided when possible because of their potential complications. (M.M.W./Ja.F.D.)

DISEASES RELATED TO WHITE CELLS

Variations in the number of white cells in humans occur normally from hour to hour, the highest counts being recorded in the afternoon and the lowest in the early morning. Temporary increases also normally occur during muscular exercise, menstruation, pregnancy, and childbirth, as well as in certain emotional states. Abnormal changes in the count, appearance, or proportion of the various white cells are indicative of pathological conditions in the body.

Leukocytosis. The condition in which white cells are present in greater numbers than normal is termed leukocytosis. It is usually caused by an increase in the number of granulocytes (especially neutrophils), some of which may be immature (myelocytes). Most often leukocytosis is the result of the presence of an infection, usually caused by pyogenic (pus-producing) organisms such as streptococcus, staphylococcus, gonococcus, pneumococcus, or meningococcus. Leukocyte counts of 12,000 to 20,000 per cubic millimetre during infections are not unusual. As the number of cells increases, the proportion of immature cells usually rises, perhaps because the demands on the leukocyte-producing tissues in the bone marrow have increased to the point at which there is an insufficient number of mature cells for delivery into the circulation. This picture of immaturity is referred to as a "shift to the left." As the infection subsides, the number of younger forms and the total white cell count decrease and ultimately return to normal. During the period of repair following an inflammatory reaction, the monocytes may increase in number, and subsequently the lymphocytes will become more numerous.

Certain types of infection are characterized from the beginning by an increase in the number of small lymphocytes unaccompanied by increases in monocytes or granulocytes. Such lymphocytosis is usually of viral origin. Moderate degrees of lymphocytosis are encountered in certain chronic infections such as tuberculosis and brucellosis.

Infectious mononucleosis

Infectious mononucleosis, caused by the Epstein–Barr virus, is associated with the appearance of unusually large lymphocytes ("atypical" lymphocytes). Those cells represent part of the complex defense mechanism against the virus, and they disappear from the blood when the attack of infectious mononucleosis subsides. Infectious mononucleosis occurs predominantly in persons from 10 to 30 years of age. It is thought to be transmitted by oral contact with exchange of saliva. Discomfort, fever, sore throat, and grippelike symptoms, together with enlargement of lymph nodes and spleen, characterize the condition. The blood serum contains an antibody (sheep cell or heterophil agglutinin) that is characteristic of the disease, but antibodies against the Epstein–Barr virus itself are more specific markers of the infection. The symptoms of this disease vary in severity in different persons, but often they are mild. Recovery takes place, as a rule, within several weeks.

Despite the immune response against the Epstein–Barr virus, the virus is never completely eliminated from the body. A small number of virus particles remain in latent form within B lymphocytes. In a normal person the immune system (the T lymphocytes in particular) keeps the latent viruses in check. If the natural responses of the immune system are suppressed (immunosuppression), however, the Epstein–Barr virus can emerge from its latent form and initiate a new round of infection. The Epstein-Barr virus is a member of the herpesvirus family, which also includes viruses that cause cold sores (herpes simplex type 1), genital ulcers (herpes simplex type 2), and shingles (herpes zoster) and the cytomegalovirus. All of these viruses show the property of latency, and in each case immunosuppression can lead to their activation. The most dramatic example of this can be found in acquired immunodeficiency syndrome (AIDS), which causes a pronounced deficiency of T lymphocytes. Patients with AIDS can develop severe and even fatal herpesvirus infections.

Monocytosis

Monocytosis, an increase in the number of monocytes in the blood, occurs in association with certain infectious processes, especially subacute bacterial endocarditis—inflammation of the lining of the heart—and malaria. Monocytosis also occurs when the bone marrow is recovering from a toxic injury.

Eosinophilia, an increase in the number of eosinophilic leukocytes, is encountered in many allergic reactions and parasitic infections. It is especially characteristic of infestation by trichina larvae, which are ingested when infected and poorly cooked pork or pork products are eaten.

Agranulocytosis

Leukopenia. The term leukopenia refers to leukocyte counts that are abnormally low (below 4,000 per cubic millimetre). Like leukocytosis, which is usually due to an increase of neutrophils (neutrophilia), leukopenia usually is due to a reduction in the number of neutrophils (neutropenia). Of itself, neutropenia causes no symptoms, but persons with neutropenia of any cause may suffer from frequent and severe bacterial infections. The term agranulocytosis refers to an acute disorder characterized by severe sore throat, fever, and marked prostration associated with extreme reduction in the number of neutrophilic granulocytes or even their complete disappearance from the blood.

Neutropenia may be due to a hypersensitivity mechanism in which a drug provokes the formation of antibodies. If the drug happens to bind to the neutrophil, the antibody accidentally destroys the neutrophil when it reacts with the drug. A second important mechanism of neutropenia is toxic damage to the bone marrow. Chemotherapeutic agents used in the treatment of cancer and leukemia often produce leukopenia (and neutropenia) by damaging the bone marrow. Drugs that cause neutropenia include pain relievers, antihistamines, tranquilizers, anticonvulsants, antimicrobial agents, sulfonamide derivatives, antithyroid drugs, and chemotherapeutic agents used for the treatment of cancer and leukemia.

Neutropenia also is associated with certain types of infections (*e.g.,* typhoid, brucellosis, measles) and with certain diseases involving the bone marrow (*e.g.,* aplastic anemia) or the spleen. In addition, sufficiently high doses of radiation will cause neutropenia, as will certain antitumour agents. Treatment includes removal of the cause of the neutropenia and the use of antibiotics appropriate for the existing infection.

Leukemia. *Nature, types, and causes of leukemia.* The term leukemia means "white blood" and arose from the discovery of extremely large numbers of white blood cells in the blood of certain persons; counts as high as 500,000 per cubic millimetre and even 1,000,000 per cubic millimetre may be found in some instances.

Leukemia is a potentially fatal disease of the blood-forming tissues that is encountered at all ages and in both sexes. There are two main varieties of leukemia, myelogenous, or granulocytic, and lymphatic. These terms refer to the types of cell that are involved. Each of these types is further subdivided into acute and chronic categories, and additional, less common varieties are recognized, as will be mentioned below.

Common causes

Filterable viruses have been shown to cause leukemia in mice, rats, cats, and cows. These animal viruses are not infectious for human cells. A human retrovirus, human T-cell lymphotropic virus (HTLV-I), has been suggested to be the cause of a type of leukemia termed T-cell leukemia. Cases of T-cell leukemia associated with HTLV-I have been found in clusters in southern Japan (Kyushu) and in the coastal region of Georgia in the United States, but sporadic cases also have been identified. Other factors, such as ionizing irradiation, contribute to the development of leukemia. Ionizing irradiation is in fact a leukemia-inducing agent in humans. Survivors of the atomic bomb in Hiroshima and Nagasaki, pioneering radiologists who used inadequately shielded apparatus, and certain patients receiving a particular kind of irradiation are known to have developed leukemia with a frequency far exceeding that of the general population. Noteworthy is the fact that almost all radiation-induced leukemia has been of the granulocytic variety. The prolonged administration of radiomimetic drugs used in cancer chemotherapy, termed alkylating agents, is also associated with an increased risk of developing leukemia. Some evidence suggests that certain industrial chemicals, notably benzene, may cause leukemia. Genetic factors may lead to an increased frequency of leukemia in certain selected instances. This is suggested by evidence that shows the higher probability for acute leukemia occurring in both identical twins if one is affected as compared to both fraternal twins under the same conditions, and the frequency of development of acute myeloblastic leukemia in children with Down's syndrome, a condition in which there is a recognized chromosome defect. Evidence for the role of trauma, hormones, infections, and psychological and other influences as factors leading to the development of leukemia is unconvincing.

Genetic studies

Advances in molecular genetics have greatly increased the understanding of leukemia. The disease seems to arise from a genetic change (mutation) in an early progenitor, or stem cell, in the bone marrow. The mutant cell passes the genetic change on to all of its progeny, thus giving rise to a clone of leukemic cells. In many cases of leukemia the mutation is detectable by analysis of the chromosomes of leukemic cells. A well-studied abnormality of this type, the Philadelphia chromosome, occurs in almost all cases of chronic myelogenous leukemia. The chromosomal aberrations affect genes that influence vital aspects of cell growth and function. These genes, the oncogenes, may themselves be mutated or their regulation may be abnormal. The entire process, beginning with the mutation, usually involves many steps that culminate in a cell with the malignant attributes of a leukemic cell. This concept of the mechanism of leukemia has influenced treatment of the disease, which aims to eliminate the mutant clone and all of its progeny by chemotherapy or bone marrow transplantation (see below).

Clinical manifestations. The terms acute and chronic refer to the duration of the untreated disease. Before the advent of modern chemotherapy, patients with acute leukemia usually died within weeks or months of the first manifestations of the disease. The life span of patients with chronic leukemia is now measured in years.

Early signs of leukemia

Leukemia primarily involves the bone marrow, but the lymph nodes and spleen may also be affected. For example, the spleen usually is enlarged in chronic myelogenous leukemia, and in chronic lymphocytic leukemia there often is enlargement of lymph nodes. Changes also take place in the leukocytes, red cells, and platelets, with consequent anemia and bleeding manifestations. The first symptoms may be weakness and an increased tendency to become fatigued because of anemia, or hemorrhages into the skin and nosebleeds, or gum bleeding due to a decrease in the number of platelets. In the acute leukemias these symptoms may be severe, the anemia may progress rapidly, and there may be fever. Chronic leukemia is more insidious in development, and the early manifestations may be overlooked until enlargement of the spleen or lymph nodes is discovered. Leukemia is recognized by examination of the blood, supplemented in most instances by examination of the bone marrow.

Acute leukemia is marked by the presence in the blood of immature cells normally not found there. In acute lymphocytic anemia (also called acute lymphoblastic leukemia), most frequently seen in children, the cells are immature forms of the lymphatic series of cells. In myeloblastic leukemia the predominant cells are the youngest recognizable precursors (myeloblasts) of the neutrophils of the blood. In a third and the least common variety, acute monocytic leukemia, the immature cells appear to be precursors of the monocytes of the blood. Myeloblastic and monocytic leukemia occur more commonly in adults and adolescents than in young children. In general, acute leukemia occurs in young persons, but no age group is exempt.

The total white cell count usually is increased but not uncommonly is normal or lower than normal (leukopenic). In such cases, abnormal immature cells may nevertheless be seen in the blood. In all forms of acute leukemia the typical cells are found in abundance in the bone marrow. "Aleukemic" leukemia refers to those instances of leukemia in which no abnormal cells are found in the blood; in these instances the leukemia is identified by examinations of the bone marrow.

Chronic granulocytic (myelogenous) leukemia is characterized by the appearance in the blood of large numbers of immature white cells of the granulocytic series in the stage following the myeloblast, namely, myelocytes. The spleen becomes enlarged, anemia develops, and the affected person may lose weight. The platelets may be normal or increased in number, abnormally low values being found only in the late stages of the disease or as an unintended result of therapy. The disease is most commonly encountered in persons between the ages of 30 and 60 years. With treatment the leukocyte count falls to normal, anemia is relieved, and the size of the spleen is greatly reduced. When the leukocyte count rises again, treatment is reinstituted. Such cycles of treatment, remission, and beginning relapse with rise of leukocyte count can be repeated many times, but a stage ultimately is reached when treatment no longer is effective. The disease then often terminates in a form resembling acute leukemia ("blastic crisis"). There is considerable variation in the duration of the disease. Although in various series mean life span has been about 3¼ years, many affected persons live in good general condition for five to 10 years and sometimes longer.

Chronic lymphocytic leukemia differs in many ways from other forms of leukemia. It occurs most often in people over 50 years of age, and its course usually is rather benign. It is mainly characterized by an increase in the number of lymphocytes in the blood, often accompanied by more or less generalized enlargement of lymph nodes and the spleen. Affected persons may carry on for many years without treatment and without any other manifestations. There may be no anemia and no loss of weight. Life span in this disease is measured in terms of five, 10, and even 15 years, occasionally even longer. Two events mark a change in the state of relative good health. One is the development of anemia, sometimes hemolytic in type, often accompanied by some decrease in the number of platelets. The other is impairment of immune mechanisms, resulting in great susceptibility to bacterial infections.

Treatment and chances of survival

Treatment differs according to the type of leukemia. Consequently, proper classification of the leukemia is the first step, once the diagnosis of leukemia has been made. Treatment of all types of leukemia reduces illness and, in acute leukemia, prolongs life.

A number of drugs are used for the treatment of leukemia and, now less frequently than before chemotherapy was available, various forms of irradiation. The therapeutic agents are all myelotoxic; *i.e.,* they injure all the cells of the bone marrow, normal cells as well as leukemic cells. Their mode of action is through direct damage to the dividing stem cell (unspecialized cell from which specialized cells develop) or by slowing or cessation of cell division. These effects may be accomplished (1) by antimetabolites, substances that interfere with the synthesis of DNA, a constituent of the chromosomes in the cell nucleus; (2) by blocking DNA strand duplication through the binding of drugs such as the nitrogen mustards with the base groups of DNA; (3) by disruption of the mitotic spindle during cell division; or (4) by interfering with the formation or functioning of RNA, which is manufactured in the cell nucleus and plays an essential role in the production of protein and in other cell functions. Drugs with different modes of action are often combined, especially in the treatment of the acute leukemias.

Much skill and experience are needed in steering the narrow path between maximum possible destruction of the leukemic cells and tolerable injury to the normal cells of the host. In the process of treatment anemia may increase; the body defenses, through the decrease in the number of neutrophils, may be impaired, and the platelets may be greatly reduced in number. Anemia can be treated with blood transfusions, and serious reductions in platelets can be met for a time with platelet transfusions.

Acute lymphoblastic leukemia is more successfully treated than are other forms of acute leukemia; prolonged remissions and even cures can be brought about in children with the disease. Certain drugs are used to bring about remission; if the remission is complete, the patient becomes well, and no signs of the disease are demonstrable in the blood or bone marrow; drugs other than those used to induce remission often are more useful in maintaining the remission than the remission-inducing drugs.

Acute myeloblastic leukemia and acute monocytic leukemia are less effectively treated by available drugs than is acute lymphoblastic leukemia. Nevertheless, new and aggressive forms of chemotherapy can induce lengthy remissions of the disease. Transplantation of normal bone marrow, following total irradiation of the patient to destroy all his normal bone marrow cells as well as the leukemic cells, has shown promise. With the advent of refined techniques of tissue typing, it is no longer necessary that the donor of the bone marrow be an identical twin. Marrow from histocompatible siblings (brother, sister) are effective. Best results have been obtained in patients younger than 35 years of age. When this form of treatment is successful, the patient's marrow is replaced by the donor's marrow. Although the treatment is arduous and complex, cures of acute lymphocytic and acute myelogenous forms of leukemia are now possible with bone marrow transplantation.

Tissue typing

Chronic myelogenous leukemia is treated with a drug, busulfan, in daily doses until the leukocyte count has returned to normal. Treatment then is interrupted until the leukocyte count has risen to about 50,000 cells per cubic millimetre, when treatment is resumed. This can be repeated many times, and thus the affected person is maintained in good health for years. Not infrequently the intervals between treatments are six months in duration or longer. Busulfan, however, like other antileukemic agents, can injure the bone marrow, and other adverse effects may occur. Other drugs and X-ray therapy also have been used but are somewhat less valuable than busulfan. In suitable young patients, bone marrow transplantation has been used successfully in chronic myelogenous leukemia.

In its early stages, chronic lymphocytic leukemia seems best untreated, as long as anemia is not present or glandular enlargement is not too troublesome. None of the current treatments are curative. The high leukocyte counts in themselves are not harmful. When there is severe anemia, however, or when the platelet count is very low and bleeding manifestations are severe, adrenocorticosteroid hormones are often given.

Lymphomas

The lymphomas are malignant tumours of lymphocytes that are usually not associated with a leukemic blood picture. Instead, enlargement of lymph nodes and/or spleen are characteristic. The lymphomas are classified into two main groups: Hodgkin's disease and non-Hodgkin's lymphoma (or lymphocytic lymphoma). Hodgkin's disease usually begins with a painless swelling of a lymph node, and it may involve lymph nodes anywhere in the body. The disease seems to begin in one lymph node and spread from there to others. Exact determination of the extent of the disease ("staging") is important in planning its treatment. Hodgkin's disease is a potentially curable lymphoma, so details of its treatment are crucial. The early stages of the disease can be cured with radiotherapy. More-advanced stages are still curable with chemotherapy, and in some patients a combination of chemotherapy and radiotherapy is used. The staging of the disease entails a thorough medical examination, bone marrow biopsy, and X rays. The latter usually include computerized axial tomography (CAT) scanning to identify enlarged lymph nodes in the interior of the body. In many cases, surgery (laparotomy) is required to obtain for examination lymph nodes from deep within the abdomen. The goal of all these staging procedures is to enable the physician to select the therapy with the maximum curative potential.

The non-Hodgkin's lymphomas constitute a heterogeneous group that arise from either B lymphocytes or T lymphocytes. They may have an indolent course, as in the nodular well-differentiated B lymphocyte lymphomas, or the tumour may be aggressive, as in the diffuse T lymphocyte forms. Unlike Hodgkin's disease, the lymphocytic lymphomas spread through the bloodstream. Therefore, the staging procedure is not as extensive as in Hodgkin's disease. Combination chemotherapy, usually given in cycles over a period of months, is effective in certain types of diffuse lymphocytic lymphomas. Prolonged remissions with eradication of the disease are difficult to achieve in the indolent nodular lymphomas.

Multiple myeloma

Another malignant disease, probably related to the above conditions, is multiple myeloma, which is characterized by a malignant overgrowth of plasma cells within the bone marrow. This severely painful disorder causes defects in the bone of the skull, the ribs, the spine, and the pelvis that ultimately result in fractures. As the bone marrow becomes more involved, anemia develops and hemorrhages occur: the number of leukocytes may be low, and abnormal myeloma or plasma cells are found in the bone marrow. This disorder is associated with a marked overproduction of immunoglobulin by the malignant plasma cell (the normal plasma cell is the source of the antibodies). The plasma cells in multiple myeloma are the progeny of a single malignant clone, and they secrete into the blood a single type of immunoglobulin moleucle—a monoclonal immunoglobulin. In some cases, a component of immunoglobulin, the light chain, may be produced in excess. These light chains appear in the urine, and in multiple myeloma they are called Bence-Jones protein. A type of chronic kidney disease often develops, probably as a result of the high concentration of Bence-Jones protein in the kidney tubules; this frequently is the ultimate cause of death. In some cases the condition remains quiescent for a time, but death is inevitable. Adrenocorticosteroid hormones and chemotherapeutic agents have been used in the treatment of multiple myeloma. (M.M.W./R.S.S.)

DISEASES RELATED TO PLATELETS AND COAGULATION PROTEINS

Bleeding disorders. In humans hemostatic failure may result from inherited or acquired defects of clotting or platelet function. The usual consequence is persistent bleeding from injuries that would normally cause little trouble. Some persons may bleed more easily than normal, perhaps even spontaneously, as a result of an increased fragility of the blood vessels. This fragility is not itself a hemostatic defect but may be associated with one.

Abnormal bleeding. Abnormal bleeding follows a pattern defined by the underlying defect. Platelet abnormalities are associated with spontaneous bleeding from the membranes of the nose, mouth, and gastrointestinal and urogenital tracts. Petechiae, purpuras, and ecchymoses are medical terms denoting different sizes of hemorrhage into the skin. In purpura there is an unexplained and abnormal fragility of the capillaries, resulting in pinpoint hemorrhages in the skin and mucous membranes.

Characteristic features

Defects in clotting. Blood-clotting defects have different clinical manifestations. The most characteristic feature of such defects, as typified by hemophilia, are recurrent, crippling hemorrhages into joints and muscles and bleeding into body cavities.

The diagnosis of hemostatic defects depends on clinical and laboratory data. Investigation of the affected person's family history is important because many of these conditions are inherited. The personal history reveals the type of bleeding and the possible effects of drugs, chemicals, allergy, infection, or dietary abnormalities. Examination reveals visible hemorrhages or vascular lesions and the effects of internal bleeding. Platelets are evaluated by counting them, examining their appearance, and testing functions such as the bleeding time, platelet adhesiveness, and platelet aggregation. Tests of clotting function include such indicators as the prothrombin time and the partial thromboplastin time. These tests allow a deficiency of any of the clotting factors to be identified and a quantitative assessment of activity of individual factors to be made. Most cases of abnormal bleeding can thus be traced to specific defects (Table 7). Precise diagnosis often permits specific and effective treatment.

Table 7: Laboratory Diagnosis of Coagulation Disorders

test	factor evaluated
prothrombin time	extrinsic pathway
partial thromboplastin time	intrinsic pathway
thrombin time	fibrinogen conversion to fibrin
fibrinogen	fibrinogen concentration
fibrin split products	fibrinolysis
bleeding time	ability to generate a platelet plug
platelet adhesiveness	adherence of platelets to glass beads
platelet aggregation	aggregation of platelets with agonists

Vascular causes of bleeding disorders. Vascular defects causing abnormal bleeding are rare. In cases of vitamin C deficiency, capillary integrity is lost, and blood oozes into the tissues. In the inherited condition hemorrhagic telangiectasia, groups of enormously dilated capillaries can be seen in the skin and mucous membranes of the mouth, nose, and gastrointestinal and respiratory tracts. The lesions appear in adult life and tend to bleed on the least provocation. The Ehlers–Danlos syndrome is a disorder of collagen synthesis in which the increased fragility of vessels causes them to be easily ruptured. Cortisone, prednisone, and other corticosteroid drugs are associated with increased capillary fragility and purpura.

Disorders of platelet number. Reduction in the number of blood platelets (thrombocytopenia) may be the result of impaired production or increased destruction of platelets. Normal platelet counts are between 150,000 and 400,000 per cubic millimetre. When the platelet count drops to 50,000 to 75,000, and particularly to 10,000 to 20,000, spontaneous bleeding may occur.

Thrombocytopenia

Thrombocytopenia is associated with such blood diseases as aplastic anemia and leukemia and is attributed to impaired production of platelets. Similarly, excessive irradiation, exposure to certain chemicals (such as benzene), or medications used in cancer chemotherapy decrease the production of platelets. In sensitive persons drugs such as quinidine provoke platelet antibodies and platelet destruction, resulting in thrombocytopenia. Thrombocytopenia also may accompany certain infections such as measles and autoimmune disorders such as systemic lupus erythematosus and idiopathic thrombocytopenic purpura.

Thrombocytopenia, if sufficiently severe, is accompanied by spontaneous bleeding from the capillaries. This causes the appearance of tiny purplish spots (petechiae) or larger black and blue areas (ecchymoses) in the skin. Bleeding occurs commonly from the nose and gums and occasionally from sites such as the urinary tract and the intestines; hemorrhage in the brain can have serious consequences.

Disorders of platelet function. Some bleeding disorders are due to abnormalities of platelet function rather than to a defect in platelet number. Glanzmann's thrombasthenia, an inherited disorder associated with a mild bleeding tendency, is due to a deficiency of glycoprotein IIb–IIIa. Bernard-Soulier syndrome, an inherited disorder associated with a pronounced bleeding tendency, is due to a deficiency of glycoprotein Ib on the platelet membrane. These platelets are unusually large and morphologically bizarre. In the gray platelet syndrome, the alpha granules are missing. Many other platelet defects exist, but they have not been fully characterized at a biochemical level.

Aspirin and platelet disorders

The most common acquired disorder of platelet function is associated with aspirin, or acetylsalicylic acid. Aspirin reacts with platelets, even when the drug is taken at low doses. This reaction impairs the ability of platelets to produce a group of chemicals known as prostaglandins. The inhibition of prostaglandin biosynthesis and the decrease in the production of thromboxane A_2 can be associated with a bleeding disorder. Other drugs have a similar effect, but aspirin is especially important because of its wide use and the sensitivity of certain persons to its action.

Coagulation disorders. Deficiencies in any of the protein factors involved in coagulation can result in hemorrhages following minor injuries. In some of these disorders, a specific deficiency is due to an inherited defect (*e.g.,* hemophilia). In others, an acquired pathological condition may be responsible for the deficiency (*e.g.,* conditions interfering with absorption of vitamin K and severe infection).

Hemophilia

Hereditary coagulation disorders. Coagulation disorders include a number of disorders that are related to defects in the clotting of blood. The best known is hemophilia, which is due to an inherited defect transmitted by the female but manifested almost exclusively in the male. The most common form of hemophilia, hemophilia A, is caused by the absence of the coagulation protein factor VIII. Of patients with hemophilia, approximately 85 percent have factor VIII deficiency. The next most common form of hemophilia, hemophilia B, is due to factor IX deficiency. Both factor VIII deficiency and factor IX deficiency have signs and symptoms that are indistinguishable. Spontaneous bleeding into joints, giving rise to severe chronic arthritis, is a common problem among patients with severe hemophilia; in addition, there is bleeding into the brain and the abdominal cavity, as well as marked bruising. In general, the greater the deficiency in either factor VIII or factor IX, the more severe the manifestations of disease. Treatment of bleeding episodes emphasizes the replacement of the missing plasma protein. In a patient with hemophilia A, factor VIII can be replaced by the infusion into a vein of plasma derived from a normal donor, the cryoprecipitate fraction of normal plasma, or a partially purified preparation of factor VIII derived from normal plasma. The peptide DDAVP (1-deamino-8-D-arginine vasopressin) is useful in treating milder forms of hemophilia A. Similarly, in a patient with hemophilia B, factor IX can be replaced by the infusion into the vein of plasma derived from a normal donor or a partially purified preparation of factor IX derived from normal plasma. Because donor plasma may be contaminated with virus particles, a number of viral diseases including hepatitis and AIDS can be transmitted by plasma protein. New methods of preparing factor VIII and factor IX, using genetic engineering techniques, should lead to the introduction of safer factor VIII and factor IX generated by recombinant DNA methods. With the current methods of medical care, hemophiliacs can live nearly normal, productive lives. Major surgery, if needed, can be accomplished by the administration of the missing protein. The life expectancy of a hemophiliac with access to quality medical care is normal.

Inherited deficiencies of other clotting proteins are also associated with clotting disorders, but these disorders are not common (Table 8).

Table 8: Common Hereditary and Acquired Coagulation Disorders

- Hereditary
  - Hemophilia A (factor VIII deficiency)
  - Hemophilia B (factor IX deficiency)
  - Hemophilia C (factor XI deficiency)
  - Von Willebrand's disease (von Willebrand factor deficiency)
- Acquired
  - Vitamin K deficiency
  - Circulating anticoagulants
  - Disseminated intravascular coagulation
  - Liver disease
  - Medications

Vitamin K deficiency

Vitamin K deficiency leads to a deficiency of the proteins that require vitamin K for their synthesis: prothrombin, factor X, factor IX, factor VII, protein C, and protein S. Vitamin K deficiency is associated with obstructive jaundice, in which the flow of bile into the bowel is interrupted. Bile is necessary for the absorption of vitamin K. Similar changes may take place when absorption of vitamin K is impaired by conditions such as chronic diarrhea or with the administration of certain antibiotics. Vitamin K deficiency also occurs in the newborn infant as hemorrhagic disease of the newborn. This form of prothrombin deficiency can be prevented by administration of vitamin K to the baby during delivery. Accidental consumption or overdoses of the medication warfarin can lead to a deficiency of the vitamin K-dependent blood-clotting proteins and a serious bleeding tendency.

Severe liver diseases such as cirrhosis and hepatitis can be associated with significant bleeding problems. The liver is the site of synthesis of the clotting proteins. When the liver tissue is diseased or replaced by other cancerous tissues, the synthesis of the clotting proteins is impaired. Often, thrombocytopenia complicates severe liver disease and adds to the bleeding tendency.

Afibrinogenemia, or hypofibrinogenemia, refers to a reduction in the amount of the clotting factor fibrinogen in the blood. This problem is seen in rare instances as an inherited disorder, but more commonly it is found as part of the syndrome of disseminated intravascular coagulation.

Von Willebrand's disease

Von Willebrand's disease is due to the lack of von Willebrand factor, a plasma protein that binds to factor VIII and promotes the interaction of platelets with the blood vessel. This disorder is transmitted as an autosomal dominant trait, and its symptoms consist mainly of bleeding through the skin and mucous membranes. Treatment involves the infusion of plasma, the cryoprecipitate fraction of plasma, or the peptide DDAVP.

Disseminated intravascular coagulation is an acquired disorder in which the blood coagulation system is activated inappropriately, and the normal regulatory mechanisms are overwhelmed. Numerous primary problems can be responsible for this activation: bacteria and bacterial products, dead or injured cells as a result of tissue injury or surgery, cells from the placenta or a dead fetus, certain forms of cancer, and venom from snake bites. Treatment involves the removal of the inciting cause of the activation of the blood coagulation system. In this disorder, platelets and blood-clotting components are consumed until a severe deficiency exists, resulting in a bleeding disorder. In addition, the fibrinolytic system—the system that dissolves clots—is also activated, leading to the destruction of fibrinogen and fibrin clots.

General treatment of bleeding disorders consists mainly of temporary replacement therapy by transfusion. In thrombocytopenia (deficiency of platelets) or thrombocytopathy (malfunctioning platelets) transfusions of normal donor platelets can be given. The clotting deficiencies are treated with plasma or plasma proteins containing the missing factor. These agents can restore hemostatic function to normal for hours or days and, with continued treatment, allow injuries to heal or complicated surgery to be performed.

Underlying cause

Thrombosis. Thrombosis, the formation of a blood clot (thrombus) that tends to plug functionally normal blood vessels, is one of the major causes of death in Western societies. In this condition, the normal tendency of platelets to form a clot at the site of a vascular injury is a contributing factor. While the process of thrombosis is not completely understood, the chief underlying cause is believed to be a lesion that destroys the normal endothelial surface of the blood vessel. Platelets tend to adhere to such lesions and eventually form masses that, when reinforced by fibrin, may completely obstruct the blood flow. This obstruction may have disastrous consequences in arteries such as the coronary or cerebral blood vessels or a major artery of a limb or organ. In veins, the local results are less apparent or perhaps not even detected; however, behind the original thrombus, the whole blood content of the vein may clot to form a large mass. This occurs most frequently in the leg veins and may be due to slowing of the blood flow during prolonged surgical operations or confinement to bed. The danger is that the thrombus may become detached (an embolus) and be swept into the pulmonary artery, causing circulatory obstruction in the lung.

Disorders of the regulation of blood coagulation may be either inherited or acquired. The hereditary deficiency of antithrombin III, protein C, protein S, and plasminogen can be associated with a thrombotic tendency; *i.e.,* the inappropriate formation of clots in vessels. The most common of the thrombotic disorders include pulmonary embolism (clots in the lung) and deep vein thrombosis (clots in the leg veins). Other predisposing causes of thrombosis are an increase in platelet numbers in the blood and the formation of functionally abnormal platelets in diseases of the bone marrow. Birth control pills, pregnancy, and many disorders may be associated with a thrombotic tendency, but the reasons for this association are not understood.

Prevention and treatment

The prevention and treatment of thrombosis aims at balancing the overactive clotting tendency and a bleeding tendency. The formation of clots may be prevented by the administration of anticoagulant drugs. Sodium warfarin inhibits the action of vitamin K and renders the blood less able to clot. Heparin facilitates the inhibition of activated clotting enzymes with the inhibitor, antithrombin III. These commonly used medications prevent the formation of clots or the enlargement of already existing clots. Streptokinase, urokinase, and tissue plasminogen activator are enzymes that dissolve clots. As agents that cause the lysis of fibrin clots, they are known collectively as fibrinolytic agents. Their administration, although attended by the possibility of bleeding complications, can restore blood flow through a vessel blocked by a clot. (B.Fu.)

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(C.L.C./S.La./M.M.W./R.S.S./E.M.B./B.Fu./Ja.F.D.)

Bohr

One of the foremost scientists of the 20th century, Niels Henrik David Bohr was the first to apply the quantum theory, which restricts the energy of a system to certain discrete values, to the problem of atomic and molecular structure. He was a guiding spirit and major contributor to the development of quantum physics.

Family and home

Early life. Bohr was born in Copenhagen on Oct. 7, 1885. His father, Christian Bohr, professor of physiology at the University of Copenhagen, was known for his work on the physical and chemical aspects of respiration. His mother, Ellen Adler Bohr, came from a wealthy Jewish family prominent in Danish banking and parliamentary circles. Bohr's scientific interests and abilities were evident early, and they were encouraged and fostered in a warm, intellectual family atmosphere. Niels's younger brother, Harald, became a brilliant mathematician.

Bohr distinguished himself at the University of Copenhagen, winning a gold medal from the Royal Danish Academy of Sciences and Letters for his theoretical analysis of and precise experiments on the vibrations of water jets as a way of determining surface tension. In 1911 he received his doctorate for a thesis on the electron theory of metals that stressed the inadequacies of classical physics for treating the behaviour of matter at the atomic level. He then went to England, intending to continue this work with Sir J.J. Thomson at Cambridge. Thomson never showed much interest in Bohr's ideas on electrons in metals, however, although he had worked on this subject in earlier years. Bohr moved to Manchester in March 1912 and joined Ernest Rutherford's group studying the structure of the atom.

At Manchester Bohr worked on the theoretical implications of the nuclear model of the atom recently proposed by Rutherford. Bohr was among the first to see the importance of the atomic number, which indicates the position of an element in the periodic table and is equal to the number of natural units of electric charge on the nuclei of its atoms. He recognized that the various physical and chemical properties of the elements depend on the electrons moving around the nuclei of their atoms and that only the atomic weight and possible radioactive behaviour are determined by the small but massive nucleus itself. Rutherford's nuclear atom was both mechanically and electromagnetically unstable, but Bohr imposed stability on it by introducing the new and not yet clarified ideas of the quantum theory being developed by Max Planck, Albert Einstein, and other physicists. Departing radically from classical physics, Bohr postulated that any atom could exist only in a discrete set of stable or stationary states, each characterized by a definite value of its energy.

By courtesy of the Nobelstiftelsen, Stockholm

Bohr.

The most impressive result of Bohr's essay at a quantum theory of the atom was the way it accounted for the series of lines observed in the spectrum of light emitted by atomic hydrogen. He was able to determine the frequencies of these spectral lines to considerable accuracy from his theory, expressing them in terms of the charge and mass of the electron and Planck's constant (the quantum of action, designated by the symbol h). To do this, Bohr also postulated that an atom would not emit radiation while it was in one of its stable states but rather only when it made a transition between states. The frequency of the radiation so emitted would be equal to the difference in energy between those states divided by Planck's constant. This meant that the atom could neither absorb nor emit radiation continuously but only in finite steps or quantum jumps. It also meant that the various frequencies of the radiation emitted by an atom were not equal to the frequencies with which the electrons moved within the atom, a bold idea that some of Bohr's contemporaries found particularly difficult to accept. The consequences of Bohr's theory, however, were confirmed by new spectroscopic measurements and other experiments.

Bohr returned to Copenhagen from Manchester during the summer of 1912, married Margrethe Nørlund, and continued to develop his new approach to the physics of the atom. The work was completed in 1913 in Copenhagen but was first published in England. In 1916, after serving as a lecturer in Copenhagen and then in Manchester, Bohr was appointed to a professorship in his native city. The university created for Bohr a new Institute of Theoretical Physics, which opened its doors in 1921; he served as director for the rest of his life.

The correspondence principle

Through the early 1920s, Bohr concentrated his efforts on two interrelated sets of problems. He tried to develop a consistent quantum theory that would replace classical mechanics and electrodynamics at the atomic level and be adequate for treating all aspects of the atomic world. He also tried to explain the structure and properties of the atoms of all the chemical elements, particularly the regularities expressed in the periodic table and the complex patterns observed in the spectra emitted by atoms. In this period of uncertain foundations, tentative theories, and doubtful models, Bohr's work was often guided by his correspondence principle. According to this principle, every transition process between stationary states as given by the quantum postulate can be "coordinated" with a corresponding harmonic component (of a single frequency) in the motion of the electrons as described by classical mechanics. As Bohr put it in 1923, "notwithstanding the fundamental departure from the ideas of the classical theories of mechanics and electrodynamics involved in these postulates, it has been possible to trace a connection between the radiation emitted by the atom and the motion of the particles which exhibits a far-reaching analogy to that claimed by the classical ideas of the origin of radiation." Indeed, in a suitable limit the frequencies calculated by the two very different methods would agree exactly.

Bohr's institute in Copenhagen soon became an international centre for work on atomic physics and the quantum theory. Even during the early years of its existence, Bohr had a series of coworkers from many lands, including H.A. Kramers from The Netherlands, George Charles de Hevesy from Hungary, Oskar Klein from Sweden, Werner Heisenberg from Germany, and John Slater from the United States. Bohr himself began to travel more widely, lecturing in many European countries and in Canada and the United States. His work on atomic theory was recognized by the Nobel Prize for Physics in 1922.

At this time, more than any of his contemporaries, Bohr stressed the tentative and symbolic nature of the atomic models that were being used, since he was convinced that even more radical changes in physics were still to come. In 1924 he was ready to consider the possibility that the conservation laws for energy and momentum did not hold exactly on the atomic level but were only valid as statistical averages. This exteme measure for avoiding the apparently paradoxical particle-like properties of light soon proved to be untenable and also unnecessary. During the next few years, a genuine quantum mechanics was created, the new synthesis that Bohr had been expecting. The new quantum mechanics required more than just a mathematical structure of calculating; it required a physical interpretation. That physical interpretation came out of the intense discussions between Bohr and the steady stream of visitors to his world capital of atomic physics, discussions on how the new mathematical description of nature was to be linked with the procedures and the results of experimental physics.

Principle of complementarity

Bohr expressed the characteristic feature of quantum physics in his principle of complementarity, which "implies the impossibility of any sharp separation between the behaviour of atomic objects and the interaction with the measuring instruments which serve to define the conditions under which the phenomena appear." As a result, "evidence obtained under different experimental conditions cannot be comprehended within a single picture, but must be regarded as complementary in the sense that only the totality of the phenomena exhausts the possible information about the objects." This interpretation of the meaning of quantum physics, which implied an altered view of the meaning of physical explanation, was gradually accepted by most physicists. The most famous and most outspoken dissenter, however, was Einstein.

Einstein greatly admired Bohr's early work, referring to it as "the highest form of musicality in the sphere of thought," but he never accepted Bohr's claim that quantum mechanics was the "rational generalization of classical physics" demanded for the understanding of atomic phenomena. Einstein and Bohr discussed the fundamental questions of physics on a number of occasions, sometimes brought together by a close mutual friend, Paul Ehrenfest, professor of theoretical physics at the University of Leiden, Neth., but they never came to basic agreement. In his account of these discussions, however, Bohr emphasized how important Einstein's challenging objections had been to the evolution of his own ideas and what a deep and lasting impression they had made on him.

During the 1930s Bohr continued to work on the epistemological problems raised by the quantum theory and also contributed to the new field of nuclear physics. His concept of the atomic nucleus, which he likened to a liquid droplet, was a key step in the understanding of many nuclear processes. In particular, it played an essential part in 1939 in the understanding of nuclear fission (the splitting of a heavy nucleus into two parts, almost equal in mass, with the release of a tremendous amount of energy).

Bohr's institute continued to be a focal point for theoretical physicists until the outbreak of World War II. The annual conferences on nuclear physics as well as formal and informal visits of varied duration brought virtually everyone concerned with quantum physics to Copenhagen at one time or another. Many of Bohr's collaborators in those years have written lovingly about the extraordinary spirit of the institute, where young scientists from many countries worked together and played together in a lighthearted mood that concealed both their absolutely serious concern with physics and the darkening world outside. "Even Bohr," wrote H.B.G. Casimir, one of the liveliest of the group, "who concentrated more intensely and had more staying power than any of us, looked for relaxation in crossword puzzles, in sports, and in facetious discussions."

Role under German occupation

Later life. When Denmark was overrun and occupied by the Germans in 1940, Bohr did what he could to maintain the work of his institute and to preserve the integrity of Danish culture against Nazi influences. In 1943, under threat of immediate arrest because of his Jewish ancestry and the anti-Nazi views he made no effort to conceal, Bohr, with his wife and some other family members, was transported to Sweden by fishing boat in the dead of night by the Danish resistance movement. A few days later the British government sent an unarmed Mosquito bomber to Sweden, and Bohr was flown to England in a dramatic flight that almost cost him his life. During the next two years, Bohr and one of his sons, Aage (who later followed his father's career as a theoretical physicist, director of the institute, and Nobel Prize winner in physics), took part in the projects for making a nuclear fission bomb. They worked in England for several months and then moved to Los Alamos, N.M., U.S., with a British research team.

Bohr's concern about the terrifying prospects for humanity posed by such atomic weapons was evident as early as 1944, when he tried to persuade British prime minister Winston Churchill and U.S. president Franklin D. Roosevelt of the need for international cooperation in dealing with these problems. Although this appeal did not succeed, Bohr continued to argue for rational, peaceful policies, advocating an "open world" in a public letter to the United Nations in 1950. Bohr was convinced that free exchange of people and ideas was necessary to achieve control of nuclear weapons. He led in promoting such efforts as the First International Conference on the Peaceful Uses of Atomic Energy, held in Geneva (1955), and in helping to create the European Council for Nuclear Research (CERN). Among his many honours, Bohr received the first U.S. Atoms for Peace Award in 1957.

In his last years Bohr tried to point out ways in which the idea of complementarity could throw light on many aspects of human life and thought. He had a major influence on several generations of physicists, deepening their approach to their science and to their lives. Bohr himself was always ready to learn, even from his youngest collaborators. He drew strength from his close personal ties with his coworkers and with his sons, his wife, and his brother. Profoundly international in spirit, Bohr was just as profoundly Danish, firmly rooted in his own culture. This was symbolized by his many public roles, particularly as president of the Royal Danish Academy from 1939 until the end of his life. Bohr died in Copenhagen on Nov. 18, 1962.

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Bolivia

The Republic of Bolivia (República de Bolivia) is a landlocked republic in central South America with an area of 424,164 square miles (1,098,581 square kilometres). The country has been landlocked since it lost its Pacific coast territory to Chile in the War of the Pacific (1879–84). Extending a maximum of 950 miles (1,503 kilometres) north–south and 800 miles east–west, Bolivia is bordered to the north and east by Brazil, to the southeast by Paraguay, to the south by Argentina, and to the southwest and west by Chile and Peru. Lake Titicaca, the second largest lake in South America and the world's highest commercially navigable lake, is shared with Peru. Although the constitutional capital is the city of Sucre, where the Supreme Court is established, the de facto capital is the city of La Paz, where the executive and legislative branches of government function.

Although only about one-third of Bolivia lies in the Andes Mountains, it is widely regarded as a highland country because the most developed and densely populated part of its territory is situated in or near the Andean system. The country itself has a rich history: it was once a part of the ancient Inca Empire and later became part of the Spanish viceroyalty of Peru, providing the mother country with immense wealth in silver. Spanish and the Indian languages Aymara and Quechua are official languages, and the overwhelming majority of the people are Roman Catholic. Bolivia, although rich in mineral resources, remains an underdeveloped country whose economic life is based principally upon agriculture and the production of raw materials, notably natural gas and tin.

(J.V.F./V.A.U.)

This article is divided into the following sections:

Physical and human geography

THE LAND

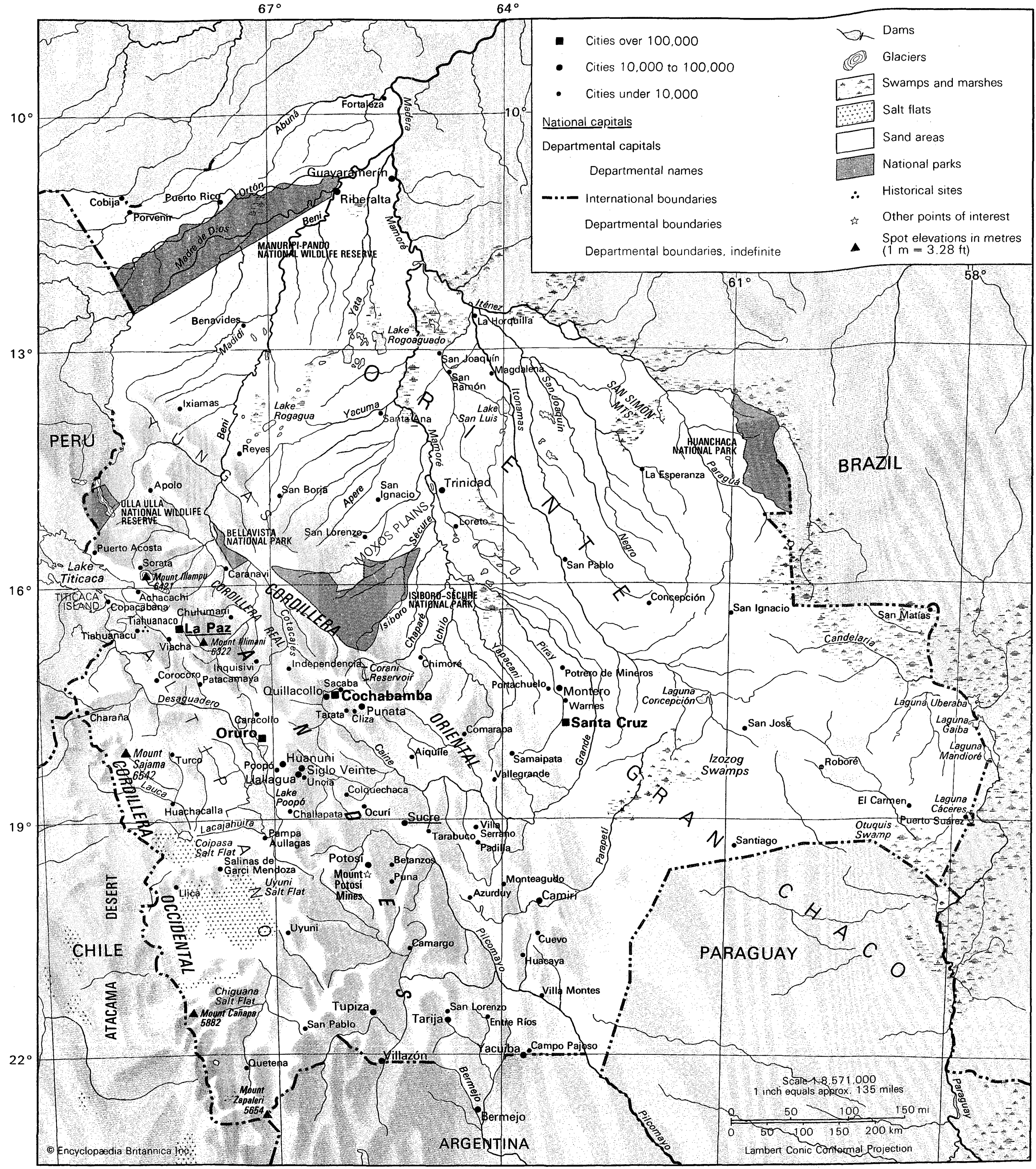
Relief. Bolivia's mountainous western region, which is one of the highest inhabited areas in the world, constitutes the country's heartland. The Andes there attain their greatest breadth and complexity. The system in Bolivia is dominated by two great parallel ranges. To the west along the border with Chile is the Cordillera Occidental, which contains numerous active volcanoes and is crowned by the republic's highest peak, Mount Sajama, at an altitude of more than 21,400 feet (6,523 metres). To the east is the Cordillera Oriental, whose spectacular northern section near La Paz is called Cordillera Real ("Royal Range"). An impressive line of snowcapped peaks, some exceeding 20,000 feet, characterize this northern section, which maintains an average elevation of more than 18,000 feet for over 200 miles. Between these ranges lies the monotonous, bleak Altiplano ("High Plateau"). The plateau is a relatively flat-floored depression about 500 miles long and 80 miles wide, lying at elevations between 12,000 and 12,500 feet (3,600 and 3,800 metres). The surface of this great plateau, composed mostly of water- and wind-borne deposits from the bordering mountains, slopes gently southward, its evenness broken by occasional hills and ridges. The margins of the Altiplano are characterized by numerous spurs and interlocking alluvial fans (accumulations of silt, gravel, and other debris that were brought down from the mountains and that have spread out in the shape of a fan), which have built up an almost continuous plain of fairly gentle grade lying at the foot of the mountains.

The Altiplano

From the high, snowcapped slopes of the Cordillera Real the descent to the eastern plains is extremely precipitous, plunging through a rainy and heavily forested belt of rugged terrain (deep valleys and gorges separated by high ridges) called the Yungas—an Aymara word roughly translated as "Warm Lands." The Yungas forms the southern end of an unbroken region that extends along the eastern Andes of Colombia, Ecuador, and Peru and continues as far south as Santa Cruz in Bolivia. South of the Yungas the Andes become much wider and are formed by a high, tilted block called the Puna, with west-facing escarpments and more gentle eastward slopes down to the plains. The region is characterized by a system of valleys and mountain basins called the Valles; these are generally larger and less confined than those in the Yungas. Lying mostly at elevations of 6,000 to 9,500 feet, they contain the so-called garden cities of Cochabamba, Sucre, and Tarija. The area is noted for its fertile valleys and rich, varied agriculture.

The Oriente

Beyond the Andes to the north and east is the Oriente region, which covers more than two-thirds of Bolivia. The vast area of the Oriente is composed of low alluvial plains (llanos), great swamps, flooded bottomlands, open savannas, and tropical forests. In the extreme south is the Bolivian Chaco, which forms part of the Gran Chaco; it is a level area that varies strikingly with the seasons. During the rainy season the Bolivian Chaco becomes a veritable swamp, but it is a hot semidesert during the remaining seven or eight months of the year. Northward from the



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Cochabamba	17 30 s 65 40 w
La Paz	15 30 s 68 00 w
Oruro	18 40 s 67 30 w
Pando	11 20 s 67 40 w
Potosí	20 40 s 67 00 w
Santa Cruz	17 30 s 61 30 w
Tarija	21 30 s 64 00 w

Cities and towns

Achacachi	16 03 s 68 43 w
Aiquile	18 10 s 65 10 w
Apolo	14 43 s 68 31 w
Azurduy	19 59 s 64 29 w
Benavides	12 38 s 67 20 w
Bermejo	22 44 s 64 21 w
Betanzos	19 34 s 65 27 w
Camargo	20 39 s 65 13 w
Camiri	20 03 s 63 31 w
Campo Pajoso	21 55 s 63 40 w
Caracollo	17 39 s 67 10 w
Caranavi	15 46 s 67 36 w
Challapata	18 54 s 66 47 w
Charaña	17 36 s 69 28 w
Chimoré	16 57 s 65 10 w
Chulumani	16 24 s 67 31 w
Cliza	17 36 s 65 56 w
Cobija	11 02 s 68 44 w
Cochabamba	17 24 s 66 09 w
Colquechaca	18 40 s 66 01 w
Comarapa	17 54 s 64 29 w
Concepción	16 15 s 62 04 w
Copacabana	16 10 s 69 05 w
Corocoro	17 12 s 68 29 w
Cuevo	20 27 s 63 32 w
El Carmen	18 49 s 58 33 w
Entre Ríos	21 32 s 64 12 w
Fortaleza	10 37 s 66 13 w
Guayaramerín	10 48 s 65 23 w
Huacaya	20 45 s 63 43 w
Huachacalla	18 45 s 68 17 w
Huanuni	18 16 s 66 51 w
Independencia	17 07 s 66 53 w
Inquisivi	16 58 s 67 10 w
Ixiamas	13 45 s 68 09 w
La Esperanza	14 34 s 62 10 w
La Horquilla	12 34 s 64 25 w
La Paz	16 30 s 68 09 w
Llallagua	18 25 s 66 38 w
Llica	19 52 s 68 16 w
Loreto	15 13 s 64 40 w
Magdalena	13 20 s 64 08 w
Monteagudo	19 49 s 63 59 w
Montero	17 20 s 63 15 w
Ocurí	18 50 s 65 50 w
Oruro	17 59 s 67 09 w
Padilla	19 19 s 64 20 w
Pampa Aullagas	19 11 s 67 05 w
Patacamaya	17 14 s 67 55 w
Poopó	18 23 s 66 59 w
Portachuelo	17 21 s 63 24 w
Porvenir	11 15 s 68 41 w
Potosí	19 35 s 65 45 w
Potrero de Mineros	17 10 s 63 14 w
Puerto Acosta	15 32 s 69 15 w
Puerto Rico	11 05 s 67 38 w
Puerto Suárez	18 57 s 57 51 w
Puna	19 46 s 65 28 w
Punata	17 33 s 65 50 w
Quetena	22 10 s 67 25 w
Quillacollo	17 26 s 66 17 w

Reyes	14 19 s	67 23 w
Riberalta	10 59 s	66 06 w
Roboré	18 20 s	59 45 w
Sacaba	17 23 s	66 02 w
Salinas de Garci Mendoza	19 38 s	67 43 w
Samaipata	18 09 s	63 52 w
San Borja	14 49 s	66 51 w
San Ignacio	14 53 s	65 36 w
San Ignacio	16 23 s	60 59 w
San Joaquín	13 04 s	64 49 w
San José	17 51 s	60 47 w
San Lorenzo	15 22 s	65 48 w
San Lorenzo	21 26 s	64 47 w
San Matías	16 22 s	58 24 w
San Pablo	15 41 s	63 15 w
San Pablo	21 41 s	66 38 w
San Ramón	13 17 s	64 43 w
Santa Cruz	17 48 s	63 10 w
Santiago	19 22 s	60 51 w
Siglo Veinte	18 22 s	66 38 w
Sorata	15 47 s	68 40 w
Sucre	19 02 s	65 17 w
Tarabuco	19 10 s	64 57 w
Tarata	17 37 s	66 01 w
Tarija	21 31 s	64 45 w
Tiahuanacu (Tiwanacu)	16 33 s	68 42 w
Trinidad	14 47 s	64 47 w
Tupiza	21 27 s	65 43 w
Turco	18 08 s	68 16 w
Uncia	18 27 s	66 37 w
Uyuni	20 28 s	66 50 w
Vallegrande	18 29 s	64 06 w
Viacha	16 39 s	68 18 w
Villa Montes	21 15 s	63 30 w
Villa Serrano	19 06 s	64 22 w
Villazón	22 06 s	65 36 w
Warnes	17 30 s	63 10 w
Yacuiba	22 02 s	63 45 w

Physical features and points of interest

Abuná, *river*	09 41 s	65 23 w
Altiplano, *upland*	18 00 s	68 00 w
Andes, *mountains*	20 00 s	67 00 w
Apere, *river*	13 44 s	65 18 w
Bellavista National Park	15 40 s	67 30 w
Beni, *river*	10 23 s	65 24 w
Bermejo, *river*	22 53 s	64 21 w
Cáceres, Laguna	18 56 s	57 48 w
Caine, *river*	18 23 s	65 21 w
Cañapa, Mount	21 28 s	68 07 w
Candelaria, *river*	17 17 s	58 39 w
Chaparé, *river*	15 58 s	64 42 w
Chiguana Salt Flat	21 09 s	68 05 w
Coipasa Salt Flat	19 26 s	68 09 w
Concepción, Laguna	17 29 s	61 25 w
Corani Reservoir	17 15 s	65 54 w
Cotacajes, *river*	16 00 s	67 01 w
Desaguadero, *river*	18 24 s	67 05 w
Gaiba, Laguna	17 45 s	57 43 w
Gran Chaco, *region*	20 00 s	61 00 w
Grande, *river*	15 51 s	64 39 w
Huanchaca National Park	14 30 s	60 30 w
Ichilo, *river*	15 57 s	64 42 w
Illampu, Mount	15 50 s	68 34 w
Illimani, Mount	16 39 s	67 48 w
Isiboro, *river*	15 28 s	65 05 w
Isiboro-Securé National Park	16 30 s	66 00 w
Iténez, *river*	11 54 s	65 01 w
Itonamas, *river*	12 28 s	64 24 w
Izozog Swamps	18 48 s	62 10 w
Lacajahuira, *river*	19 21 s	67 54 w
Lauca, *river*	19 10 s	68 10 w
Madera, *river*	10 30 s	65 30 w
Madidi, *river*	12 32 s	66 52 w
Madre de Dios, *river*	10 58 s	66 09 w
Mamoré, *river*	10 23 s	65 23 w
Mandioré, Laguna	18 08 s	57 33 w
Manuripi-Pando National Wildlife Reserve	11 30 s	67 30 w
Mount Potosí Mines	19 37 s	65 45 w
Moxos Plains	15 15 s	65 30 w
Negro, *river*	14 11 s	63 07 w
Occidental, Cordillera, *mountains*	19 00 s	68 30 w
Oriental, Cordillera	16 00 s	67 30 w
Oriente, *region*	15 00 s	64 00 w
Ortón, *river*	10 50 s	66 04 w
Otuquis Swamp	19 20 s	58 30 w
Paraguá, *river*	13 34 s	61 53 w
Parapetí, *river*	18 58 s	62 21 w
Pilcomayo, *river*	22 17 s	62 40 w
Piray, *river*	16 32 s	63 45 w
Poopó, Lake	18 45 s	67 07 w
Real, Cordillera	17 00 s	67 10 w
Rogagua, Lake	13 43 s	66 54 w
Rogoaguado, Lake	12 52 s	65 35 w
Sajama, Mount	18 06 s	68 54 w
San Joaquín, *river*	13 08 s	63 41 w
San Luis, Lake	13 45 s	64 00 w
San Simón Mountains	13 40 s	62 16 w
Sécure, *river*	15 10 s	64 52 w
Sol, Island of, see Titicaca Island		
Tiahuanaco, *historical site*	16 33 s	68 41 w
Titicaca Island (Sol, Island of)	16 01 s	69 10 w
Titicaca, Lake	15 48 s	69 24 w
Uberaba, Laguna	17 30 s	57 45 w
Ulla Ulla National Wildlife Reserve	15 00 s	69 15 w
Uyuni Salt Flat	20 20 s	67 42 w
Yacuma, *river*	13 38 s	65 23 w
Yapacani, *river*	16 00 s	64 25 w
Yata, *river*	10 29 s	65 26 w
Yungas, *region*	14 30 s	68 00 w
Zapaleri, Mount	22 49 s	67 11 w

Chaco the relief of the Santa Cruz area is somewhat more varied, exhibiting a gentle downward slope to the north. The Oriente includes the northern departments of Beni and Pando, where the low plains are covered by savanna and, in the far north, by expanses of tropical rain forest. Much of the Beni department suffers from extensive flooding beginning in March or April, toward the end of the summer rainy season.

Drainage. The rivers of Bolivia belong to three distinct systems—the Amazon system in the northeast, the Río de la Plata system in the extreme southeast, and the Lake Titicaca system in the Altiplano. The eastern lowlands of Bolivia have many lakes, most of them little known except through observation from aircraft.

The great swampy plains along the Beni and Mamoré rivers, which are headwaters of the Amazon, contain several lakes and lagoons, some of them large, such as Lake Rogoaguado. The Amazon headwaters cut deeply into the Andes; even La Paz in the far west is in the Amazon drainage basin.

In the vicinity of the Paraguay River (which runs parallel to Bolivia's far eastern border and is part of the La Plata Basin) there are several shallow lakes (partly produced by obstructed outlets such as the Bahía Negra), the largest being Cáceres, Mandioré, Gaiba, and Uberaba. North of these are the great Xarayes swamps. This region, like that in the northeast, is subject to widespread flooding during summer.

The third drainage system is that of the Altiplano, which constitutes the largest region of inland drainage in South America. One of the most elevated of all inland basins, it consists of Lake Titicaca near the north end; the Desaguadero River, Titicaca's outlet to the southeast; Lake Poopó, into which the Desaguadero flows; the Lacajahuira River, which drains westward into the Coipasa Salt Flat; and the great Uyuni Salt Flat, independent of the rest of the system but receiving the waters of an extensive though generally arid area at the south end. Into this system enter many short streams from the neighbouring heights. Having no outlet to the sea, the water of the Altiplano system is wholly absorbed by the dry soil and by excessive evaporation.

Lakes Titicaca and Poopó

Lake Titicaca itself covers 3,200 square miles (8,300 square kilometres) and is South America's largest inland lake. Situated on the Bolivian-Peruvian border at an elevation of 12,500 feet (3,810 metres), it is about 120 miles long and a maximum of 50 miles wide (190 by 80 kilometres); it reaches a maximum depth of more than 900 feet. Many islands dot the lake's surface. The basin's drainage system maintains Titicaca as a largely freshwater lake despite its high evaporation rate. Lake Poopó, which is very salty, is quite different in character from Lake Titicaca. Occupying a very shallow depression in the plateau, only a few feet below the general level of the surrounding land, Lake Poopó is rarely more than 10 feet deep at its normal level. When its waters are low, it covers an area of about 1,000 square miles; the surrounding land is so flat, however, that at high water the lake reaches sometimes almost to Oruro to the north, fully 30 miles from the low-water shore. The Lacajahuira River, the only visible outlet of Lake Poopó, disappears underground for part of its course and empties into the Coipasa Salt Flat, which at high water covers about the same area that Lake Poopó does at low water; it usually consists, however, only of wide, marshy, salt-encrusted wastes, with a small permanent body of water in the lowest part. There is no outlet. The Uyuni Salt Flat, which lies to the south of the Coipasa Salt Flat, is similar, but it is much larger. Covering about 4,000 square miles, it consists of a great windswept expanse that is totally arid.

Soils. The thin soils of the Altiplano—mainly clays, sands, and gravels—are dry and loosely consolidated; slopes that are exposed to strong winds or storm water are severely eroded. Soils to the south of the plateau are highly saline, but, in the north, rich lake silts border Titicaca, where the waters have receded over the centuries to expose the old lake floor. It is thought that Tiahuanaco, the important pre-Inca site that is now about 10 miles from the southern shore of Lake Titicaca, was once a lakeside port. In the Yungas the soils on the steep valley sides erode rapidly wherever forest is cleared and the slopes are not carefully terraced. The wider, sunnier basins in the Valles region, particularly around Cochabamba, contain deeper, more fertile soils that respond well to irrigation. East of the Andes, the soils over much of the Oriente are poor, lacking in nutrients, and alternately parched or waterlogged.

Climate. In spite of the fact that Bolivia lies wholly within the tropics, it possesses every gradation of temperature from that of the equatorial lowlands to arctic cold. In the Andes, contrasts in temperature and rainfall depend more on altitude than on distance from the Equator. Cold winds sweep the Altiplano with little variation between summer and winter, and rainfall is slight, coming mostly in the form of summer thunderstorms during December and January. Average temperatures range between 45°

and 52° F (7° and 11° C), but winter temperatures fall below freezing, and nights are cold throughout the year. In the north, however, Lake Titicaca has an important moderating influence, and in bright sunshine winter temperatures may reach as high as 70° F (21° C). Cloudless skies and astonishingly clear air bring distant peaks sharply into focus, providing hauntingly beautiful vistas across the Altiplano.

The cloud-filled Yungas valleys

In stark contrast, clouds of moist air from Amazonia to the northeast fill the valleys of the Yungas throughout the year, leaving the humid atmosphere rich with the smell of vegetation. The mean annual temperatures vary between 60° and 68° F (16° and 19° C); precipitation ranges up to 53 inches (1,350 millimetres) annually, occurring throughout the year but heaviest between December and February. The Valles have brighter conditions and less precipitation than the Yungas, as well as somewhat warmer temperatures.

On the low plains of the Oriente the climate is hot, averaging 73° to 77° F (23° to 25° C) in the south and up to 80° F (27° C) in the north. Occasional cold winds called *surazos* blow from the south; laden with sand and dust, they last for a few days, lowering temperatures abruptly. Annual rainfall ranges from about 40 inches in the south to 70 inches or more in the far north, with a pronounced summer maximum. (G.McC.McB./J.V.F.)

Plant life. Huge expanses of the southern Altiplano are saline and barren, but coarse bunchgrass (ichu; Spanish: *paja brava*) is common in the north, where it is grazed by llamas. Tola, a tough wind-resistant shrub, and mosslike cushions of yareta, both widely used for fuel, are well distributed, along with cactus scrub. Totora rushes, which grow on the shores of Lake Titicaca, are used for thatching and for making the Indian boats called balsas. The Altiplano is devoid of native trees, but around Lake Titicaca and in sheltered valleys the eucalyptus has been successfully introduced.

The Yungas is clad in luxuriant mountain jungle that includes an enormous variety of tropical hardwoods, dyewoods, medicinal and aromatic plants, and fruit trees. Characteristic trees include the green pine; aliso, a shrub-like tree; laurel; cedar; *tarco,* a shade tree producing masses of yellow-white flowers; and *saúco,* which yields fruit used to make medicinal syrups. The cinchona, or quina tree, from which quinine is made, and the coca shrub, the source of cocaine, are also indigenous to this region. In the Valles region to the south there is a general covering of drought-resistant grasses, shrubs, and small trees. In the southern foothill zone a strip of deciduous forest includes such tree species as the walnut and quebracho, the latter being a source of tannin and timber. At lower elevations in the Oriente, vegetation is strongly controlled by the degree of waterlogging that occurs and by the length of the dry season. In the south the Chaco is scrub-covered, with scattered stands of quebracho giving way northward to a transitional region of semideciduous tropical forest. Farther north in the Oriente, grass, palm, and swamp savannas extend into Beni. There, strips of tropical forest line the riverbanks, while more continuous forest appears along the Andean foothills and in eastern Bolivia. True Amazonian rain forest (selva) occurs only in the far north, in the department of Pando and adjacent areas. Among the thousands of different tree species are rubber, Brazil nuts, and mahogany.

Animal life. Bolivian highland animal life is distinguished by the presence of members of the camel family—the llama, alpaca, guanaco, and vicuña—all native to the Andes. The llama and alpaca are domesticated varieties of the wild guanaco, which survives in the mountains. The llama, the largest animal on the Altiplano and seldom seen below heights of 7,500 feet, is the traditional beast of burden and is a source of meat, wool, leather, tallow, and fuel (in the form of dried dung). The smaller alpaca is reared for its soft wool, although the wild, now rare vicuña produces an even silkier type of wool. Highland rodents include the chinchilla, the viscacha (a burrower), the mara (a long-legged, long-eared rodent), and the cavy (a guinea pig that is bred for its meat and often kept as a pet). The Andean condor (a New World vulture and the largest flying bird in the Americas) roosts and breeds at heights between 10,000 and 16,000 feet but descends to sea level in search of food on the west side of the Andes. Many smaller birds and waterfowl, including grebes, coots, cormorants, ducks, geese, and gulls, live around Lake Titicaca, while large flocks of flamingos appear farther south around Lake Poopó.

The llama and other highland animals

The rivers of the eastern plains, most of which belong to the Amazon system, have an abundance of fish, and there are numerous frogs, toads, and lizards, along with myriad forms of insect life. The armadillo, the anteater, the peccary (wild pig), the puma, and the marsh deer all inhabit the plains, as does the capybara, the largest rodent in the world, and the rhea, a flightless bird that bears a remarkable resemblance to the ostrich, although it is much smaller.

The rich animal life of the northern forests includes such mammals as the jaguar (the largest of the American cats), the sloth, the monkey, and the tapir; the largest of the numerous reptiles is the caiman, a member of the alligator family, and among the many fish species is the dreaded caribe (piranha). Varieties of snakes include constrictors and venomous species such as the fer-de-lance and the bushmaster. Many brightly coloured birds, such as parrots and toucans, inhabit the forests, seldom descending to the forest floor, and high in the sky above them may be seen the king vulture and the black vulture, gliding in search of carrion.

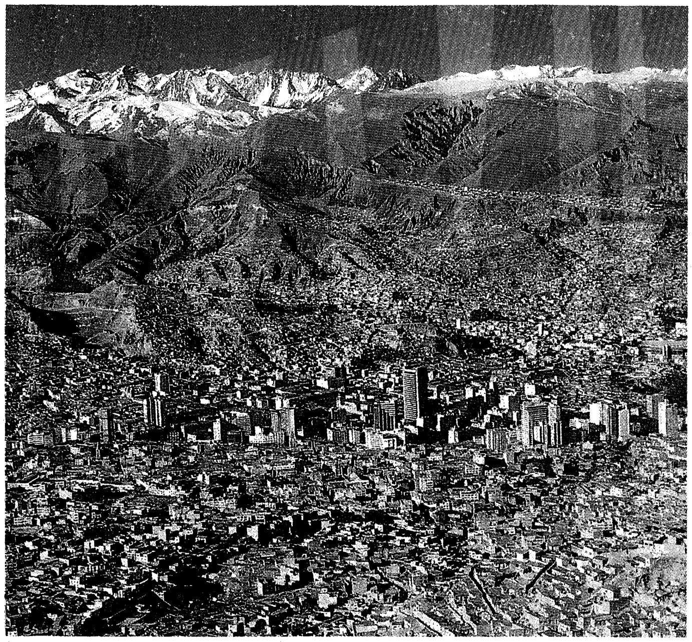
Settlement patterns. The three principal regions of settlement are the Altiplano, the Valles, and the Santa Cruz region of the Oriente.

The Altiplano, which constitutes nearly one-tenth of the total land area of Bolivia, is high, windswept, and cold. The Incas, however, found that this portion of the tropics could be cultivated and that it was healthier and more invigorating than the hot, wet plains. The northern Altiplano has remained Bolivia's most densely populated region; the cities of La Paz and Oruro are located there, as are many small towns and villages.

Much farming is still of the subsistence type, but there have been dramatic changes since the National Revolution of 1952. Until the early 1950s the land was held primarily in the form of large estates, most of which dated to the days of the Spanish conquistadores, although some land was held communally by the Indians. Following the 1953 Agrarian Reform Act, the large estates were broken up and plots of land given to the Indians, who are now called campesinos (peasants) rather than *indios,* a term which has derogatory implications. Despite initial confusion caused by the sheer speed of the reform, the reduced agricultural production, and the disruption to marketing, there has been an infusion of fresh spirit and purpose among Bolivia's new peasant landowners since the 1950s. One development has been the growth of new roadside market towns on the northern Altiplano, where peasants sell their farm surpluses and a wide range of other goods. These are carried to market on foot, by bicycle, or by truck from the valleys of the Yungas. Other campesinos bring wares from La Paz. The National Revolution also introduced a new domestic colonization program that was designed to increase food production and encourage the campesinos to leave the most densely populated parts of the Altiplano and the Valles. Three areas were selected for new settlement: the Yungas northeast of La Paz, the Chaparé foothills below Cochabamba, and the plains of the Oriente around Santa Cruz. For the last of these regions, the opening in 1954 of the paved Cochabamba–Santa Cruz highway was of crucial importance because it relieved centuries of isolation between the Andes and the plains. Within 25 years about 65,000 families had settled in these pioneer zones. The domestic colonization program, however, failed to relieve significantly the population pressure in the Andean highlands, where in the same period the population increase was nearly 10 times that of the number of eastern settlers.

Cities

The city of La Paz is the largest and most important in Bolivia. It lies in a large, spectacular canyon cut below the surface of the Altiplano, a sheltered location selected by the Spaniards in 1548 on the main silver route to the



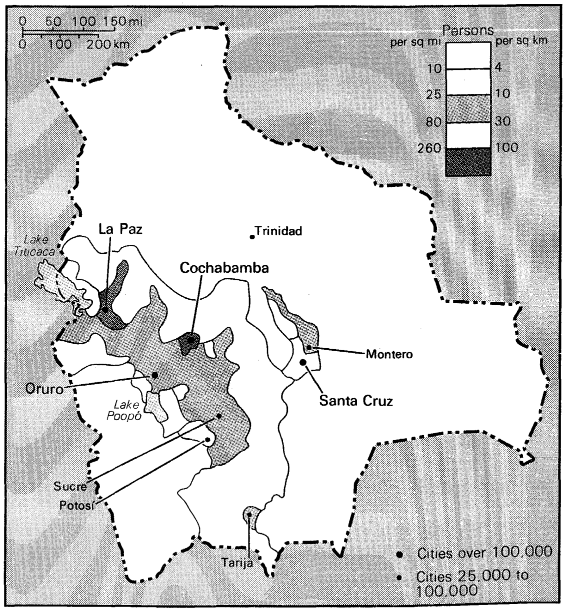
La Paz with the Cordillera Real of the Andes in the background.

Robert Frerck—Odyssey Productions

Pacific coast. Little colonial architecture survives there. La Paz grew rapidly in the late 19th and early 20th centuries as the railway centre and de facto capital of the country. The industrial and lower-income areas of the city are located high up on the valley sides, the commercial district is at the middle level, while the middle-class residential areas are at the lower levels.

The other cities of the Altiplano—Oruro, Uyuni, and Tupiza—are also railway towns and are connected with the mining industry. Potosí, east of the Altiplano, merits special attention. It was established in 1545 on the slopes of the mountain of Potosí (Cerro Rico), which was the richest source of silver found by the Spaniards. Potosí had about 160,000 inhabitants in the mid-17th century, when it was the largest city in the Americas. Even now, at more than 13,000 feet, Potosí is the highest city of its size in the world. It is one of the few cities in Bolivia to retain its architectural personality through the years.

The most important cities in the Valles were founded in the 16th century and include Cochabamba, Sucre, and Tarija. All three are surrounded by farms, fruit orchards, and dairying land. Cochabamba is the largest, busiest, and most accessible of the cities; Tarija is the most isolated—its mountain roads are tortuous, and the city has never been linked to Bolivia's rail system. Unlike conditions on the Altiplano, the climate is mild, and the lower elevations are more agreeable.



Population density of Bolivia.

The Oriente is the largest and most sparsely populated region. Santa Cruz, the only major city, lies close to the Andean foothills but is very much a city of the plains. Since the mid-1950s it has been the fastest-growing centre of agricultural colonization in Bolivia and the centre of oil and natural gas production. By the 1970s Santa Cruz had overtaken Cochabamba to become Bolivia's second largest city—a unique example of a long-isolated town in the Oriente overtaking a major Andean centre. Trinidad is the main town in the heart of the remote, sprawling, cattle-ranching region of the Beni, while farther north in the Oriente a few small towns survive along the riverbanks of the rain forest.

THE PEOPLE

Ethnic and linguistic groups. The population of Bolivia consists of three groups—the Indians, the mestizos (of mixed Indian and Spanish descent), and the descendants of the Spaniards. After four centuries of intermixing it is, however, virtually impossible to measure accurately the percentage of each, although Indians still form some 60 to 70 percent of the total, the largest group being the Quechua.

The Indians

The Indians are mainly composed of two distinct groups—those living on the northern Altiplano, who speak the guttural Aymara language, and those who speak Quechua, the language of the Incas. The Quechua are more widely distributed through the Andes, especially in the Valles. Remnants of plains and forest Indians survive in the Oriente. The great majority of the Indians are farmers, miners, and factory or construction workers. Aymara and Quechua have been added to Spanish as the official languages of Bolivia, but increasing numbers of Indians, particularly in the cities, market towns, and new colonies, speak or understand Spanish.

The mestizos

The mestizos are well represented in the offices, trades, and small businesses in the cities. The traditional minority—those of Spanish descent—have long formed the local aristocracy in small towns and rural areas. Their influence remains, although it has diminished since the National Revolution of 1952.

Few foreigners have emigrated to, or reside in, Bolivia. Small numbers of Germans arrived in the late 19th and early 20th centuries, however, and established themselves with notable success as business agents and entrepreneurs, shopkeepers, and accountants. Japanese and Okinawan farmers have been among the most successful colonists in the Santa Cruz region. Arriving in the late 1950s and 1960s as a relatively small but skilled group of pioneers, they have made an important contribution to the economy.

Religious groups. The Roman Catholic religion has the adherence of nearly 95 percent of the population. At the head of the church hierarchy in Bolivia is a primate cardinalship, located in Sucre. The churches and cathedrals, most of which were built during colonial times, constitute a national architectural treasure. They are generally in the extravagantly ornamented baroque style, although some are in Renaissance (*e.g.,* the cathedral of La Paz) or in later styles. Since the 1940s the Roman Catholic Church has ventured from an almost exclusively ceremonial role into the fields of social aid and education.

Pre-Columbian religious survival

In the Indian communities of the Altiplano, some of the characteristics of pantheistic pre-Columbian religion have survived. Its deities included the Sun God, legendary creator of the first Inca emperor Manco Capac and his sister-wife Mama Ocllo on the Island of the Sun in Lake Titicaca. The Roman Catholic religion has through the centuries accepted some aspects of the Indian religions by assimilating them into the religious life of these communities. There are also various Protestant denominations, as well as a small Jewish community. Freedom of religion is guaranteed by the Constitution.

Demography. At the beginning of the 20th century the population of Bolivia was about 1,800,000, and after 25 years of slow population growth the number had increased to only about 2,300,000. Between 1925 and 1950 the population grew at a slightly accelerated rate (despite the losses of the Chaco War), increasing by about 750,000. The 1950 population of more than 3,000,000 increased dramatically by more than 2,250,000 during the next 25 years, owing primarily to reductions in the death rate and a consistently high birth rate. By the late 20th century the 1950 population had more than doubled.

Urbanization in Bolivia parallels the pattern of population growth. At the beginning of the 20th century the urban population of Bolivia was less than one-tenth of the total, but by 1950 the urban population had more than doubled. By the late 20th century the urban population had grown to about one-half of the total population.

THE ECONOMY

Bolivia is well endowed with natural resources, but high costs of production, lack of investment, inadequate internal transport, and a landlocked location have limited its development. Average income is low, and Bolivia remains one of the poorest nations in South America.

Agrarian reform

The revolutionary program of 1952–53 included immediate agrarian reform, based upon the breakup of the large estates, and nationalization of the mines. This resulted in an initial decrease in agricultural production, a disastrous drop in mineral output, and wage increases. In a period of political turmoil, government attempts to satisfy the new labour unions delayed urgently needed reduction of the numbers employed in the mines and the promotion of greater efficiency in many other sectors of the economy. Thus, despite the long-overdue political and social reforms embodied in the revolution, the rate of national economic growth, at best, remained extremely low. The wildly fluctuating economy declined severely in the late 20th century mainly as a result of falling world prices for tin, bad harvests, debt repayments, and soaring inflation. Although Bolivia has received emergency foreign aid, technical assistance, and World Bank loans over the years, daunting problems that have no quick or easy solutions continue to face the government.

Tin

Resources. Bolivia's mineral deposits are its most valuable natural resource. The country is a major producer of tin; it has reserves of zinc, antimony, tungsten (wolfram), silver, lead, and copper, and there are small quantities of gold. Although tin dominates metal production, Bolivia is a high-cost producer compared with Southeast Asia and is thus exceptionally vulnerable to changes in world demand. The state mining corporation, Comibol, reported heavy losses and declining production in the 1980s when, faced with a world surplus, more than two-thirds of the mining work force became unemployed.

The oil and natural-gas industry

Development of Bolivia's petroleum resources dates from 1920, when the Standard Oil Company of New Jersey acquired a concession to explore and exploit the Andean foothill zone in southeastern Bolivia. A series of small oil fields was discovered there, but Standard Oil's operation was expropriated in 1937 to form the nationalized Yacimientos Petrolíferos Fiscales Bolivianos (YPFB). In the mid-1950s U.S. companies were again encouraged to resume operations, and in 1956 the Bolivian Gulf Oil Company (a branch of Gulf Oil Corporation) began a decade of successful oil and natural gas strikes in the Santa Cruz region. In 1966 Gulf began exporting oil to southern California, via the YPFB pipeline to the Pacific port of Arica (Chile), as well as boosting the Bolivian state company's sales to the domestic market and to Argentina. Political uncertainties have disrupted this industry, however, and in 1969 Bolivia nationalized Gulf Oil. Although Bolivia once again encouraged foreign oil companies in 1972, production continued to fall through lack of investment and failure to replace depleted wells. Soaring domestic consumption forced some oil importation in the late 20th century and led to a new wave of combined Bolivian and foreign oil exploration.

Natural gas production has been more successful, and with falling world markets for tin in the late 20th century, natural gas became Bolivia's most important legal export in value by the mid-1980s, accounting for more than half of official total earnings, Argentina being the principal destination. Gas fields are concentrated in the Santa Cruz region, with good reserves. Bolivia is rich in other natural resources, particularly hydroelectric potential, but exploitation has been minimal.

Agriculture. About half of the working population is engaged in agriculture (including small numbers in hunting, forestry, and fishing), but farming accounts for only about 27 percent of the gross domestic product. Although peasant markets have grown in the northern cities and roadside towns on the Altiplano and around Cochabamba, subsistence farming remains widespread in the Andes.

Highland farming

Potatoes and oca (also an edible tuber) are indigenous staple crops on the northern Altiplano, where they are eaten mainly in the dehydrated forms known as *chuño* or *tunta*. The two important grains that ripen at this altitude, both highly nutritious, are *quinoa* and *cañahu*. Llamas and alpacas are raised in the Andean region and serve a variety of agricultural functions, although the use of the llama as a pack animal has decreased with the growth of truck transport. (See above *Animal life*.)

In the Yungas excellent coffee, cacao, citrus fruits, bananas, avocados, pineapples, mangoes, papayas, melons, chili peppers, sweet potatoes (yams), and cassava (manioc) are among the astonishing variety of crops produced. Development is restricted, however, by poor roads and limited markets. Only coca (see below *Trade*) has a strong international market and a high return on production. In the warm, agreeable climate of the Valles, corn (maize), wheat, barley, alfalfa, grapes, peaches, and vegetables are grown, and sheep and dairy cattle are raised; the area, which is characterized as the garden of Bolivia, is a fertile region that has the potential, with more systematic irrigation, for greatly extended cultivation and increased yields.

In the Oriente around Santa Cruz, sugarcane, rice (dry and paddy), cotton, and beef cattle are significant products, raised mostly for the domestic market. There are also large cattle ranches in the Beni. Tropical hardwoods grow in the quebracho forests of the Chaco and among the tree species of the northern rain forest, but the great distance to market, and competition from more accessible locations outside Bolivia, limit exploitation.

Industry. The manufacturing sector has grown since the 1950s but remains small, despite some stimulus from Bolivia's membership in the Andean Group, a regional trade organization. Historically, mineral processing (including oil refining) and the preparation of agricultural products have dominated Bolivian industry. In the 1970s manufacturing employed about 10 percent of the working population, but this proportion declined somewhat in the 1980s.

Tin mining

Bolivian tin is extracted mostly by shaft mining, which is often very deep and is carried on in difficult, remote sections of the eastern cordilleras at great altitude. The ores are low-grade, often varied in content, and difficult to refine. By the late 20th century, ores with as little as 1 percent tin, or less, were being worked. Bolivian smelters under normal economic conditions produce concentrate for export, the more complex ores being smelted overseas. Food industries include flour milling, dairying, sugar refining, brewing, and alcohol distilling. Other manufactures consist of machinery, textiles, shoes, furniture, glass, bricks, cement, paper, and a wide range of small goods designed to meet the needs of a limited home market with little purchasing power. About two-thirds of Bolivia's manufacturing industry is found in or near La Paz; much of the remainder is located in the department of Cochabamba. There is severe competition from imported Brazilian, Argentine, and Peruvian manufactures, as well as from smuggled goods—the latter being mostly consumer items from neighbouring countries and the United States that can be sold cheaply because they avoid duty.

Bolivia has abundant sources of energy, particularly hydroelectric potential, but per capita consumption remains low. The National Electric Company (Empresa Nacional de Electricidad) mainly supplies the major towns and cities and the mines. Hydroelectric power stations serve La Paz, Cochabamba, and Oruro, the largest single plant being at Corani in the department of Cochabamba. Santa Cruz, Sucre, Tarija, and Trinidad are provided with thermal electric power from oil. Small generators, public and private, supply the more isolated centres in the Oriente.

Tourism does not yet contribute greatly to the economy, but encouraging efforts have been made to develop its potential in selected areas and to improve hotel accommodations, especially in La Paz. By the late 20th century Bolivia had been added to an increasingly popular grand tour of South America—a package tour of continental highlights that attracts visitors from the United States, Europe, and Japan. The main interest focuses on Lake Titicaca and its surroundings: Inca ruins on the Island of the Sun, pre-Inca ruins at Tiahuanaco, lake fishermen, Indian life on the Altiplano, and the city of La Paz itself. A side trip by road over the Cordillera Real and down into the Yungas jungles closest to La Paz is thrilling and popular, providing within a few hours some of the most dramatic scenic and climatic contrasts in the Andes.

Exports

Trade. The export of metals (mostly tin but also including zinc, silver, and tungsten) has traditionally dominated Bolivia's trade. With the collapse of the world market in tin in the 1980s, natural gas became the leading export; together, minerals and natural gas account for more than 80 percent of Bolivia's legitimate export trade. Agricultural exports include coffee, sugar, and timber, together with small quantities of wild rubber, Brazil nuts, hides, and skins. Manufactured products constitute the largest segment of total imports; machinery and equipment for industry and transport are among the main items. Raw materials, consumer goods, and food products are other major import categories. The largest overall trading partners are Argentina and the United States, but considerable trade is also carried on with other South American countries, the United Kingdom, West Germany, and Japan.

Illegal trade in cocaine

Illegal trade in cocaine has become a significant element in the Bolivian economy. The leaves of the indigenous coca shrub have been chewed by the Andean Indians for centuries as a relief against cold and fatigue. Small quantities of coca (from the Quechua *kúka*) have also been legally exported for many years for medicinal purposes. The unprecedented expansion of coca cultivation in the Yungas, and especially in the Chaparé region northeast of Cochabamba, began in the 1960s with the sudden growth in the illegal international market for cocaine. As demand soared in North America and Europe in the 1970s and 1980s, Bolivian peasant farmers soon found that no other crop could compete with coca for profitability. It became the ideal cash crop—easy to grow, valuable, non-perishable, and easy to transport, whether in dried leaf form or as a concentrate ready to be smuggled out of the country from airstrips scattered across the Oriente. By the late 20th century it was estimated that one-third of the world's coca was grown in Bolivia, one-fourth in the Chaparé region alone. Attempts by the government to introduce crop substitution or a voluntary reduction by the peasants of their coca acreage had met with little success. Indeed, the area under coca continued to increase. By the 1980s an estimated $5,000,000,000 worth of cocaine was leaving Bolivia annually from centres around Cochabamba and from remote parts of the Oriente. Despite its illegal nature the drug trafficking has provided a huge addition to the country's gross national product. Bringing needed revenue to both the government and the farmer and untold riches to the illegal dealers, the cocaine trade has become almost impossible to suppress.

Transportation. Bolivia has been hindered in its growth both by its landlocked location and by a difficult internal geography of steep mountains and seasonally flooded plains. The main rail system, in the west, was built mostly between the 1890s and the 1920s and links the major Andean cities and mines with the Pacific ports of Antofagasta and Arica (Chile) and Matarani (Peru), the latter line being connected by Lake Titicaca steamers. Isolated from this network, Santa Cruz, forming the nucleus of an eastern system, was linked by rail to Corumbá in Brazil and to Argentina via Yacuiba during the 1950s; neither line carries much traffic, however.

Road transport has developed dramatically in highland Bolivia and around Santa Cruz since the mid-1950s, and there are connecting paved highways for La Paz–Oruro, Cochabamba–Santa Cruz, Santa Cruz–Montero, and other major colonization centres in the Santa Cruz region. The Pan-American Highway cuts across highland Bolivia between Peru and Argentina, and bus and truck services on unpaved roads connect numerous Andean towns. Journeys are slow and often hazardous. Trails are used in parts of the Oriente after the floods subside; there is no road network. The immense river system sweeps in great loops over the plains, but local traffic is limited. Only in the northern rain forests are small riverboats important in regional transportation.

Air transport

Air transport is the only fast link between Bolivia's major cities and the primary means by which the isolated settlements in the Oriente are connected to the rest of the country. The national airline—Lloyd Aéreo Boliviano (LAB)—was founded by a small group of German businessmen in 1925, and in the second half of the 20th century it has played an indispensable political role in helping to identify the plains and the eastern border regions as Bolivian territory. Newspapers and information are quickly distributed by air services, keeping far-flung clusters of population informed of national affairs. LAB flies international routes to South American capitals, as well as to other cities such as São Paulo, Rio de Janeiro and Manaus (Brazil), Panama City, and Miami, Fla.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Bolivia was declared independent in 1825 and adopted its first constitution in 1826. Despite revisions and numerous coups d'état, the state has retained a unitary system of government, whether elected or under military dictatorship, the latter having held sway for much of Bolivia's history. Under the conditions of the 1947 constitution, executive power is vested in a president, who is directly elected by popular vote for a term of four years. If no candidate receives an absolute majority of votes, the National Congress must select the president from among the three leading contenders. The president cannot be immediately reelected. The bicameral legislature consists of a Senate and a Chamber of Deputies, whose members are directly elected for four-year terms. The judicial system is headed by a Supreme Court of 12 members, who are appointed by Congress for 10-year terms.

The electoral system

The country is divided into nine departments, each of which is headed by a prefect appointed by the Bolivian president. Departments are subdivided into provinces administered by subprefects, and provinces are subdivided into cantons administered by corregidors.

Following the National Revolution of 1952, universal suffrage was introduced, and all men and women became eligible to vote for the first time in Bolivia's history. Previously, though open to both sexes, the franchise had literacy and property requirements and was thus restricted to a tiny proportion of the population. The nation's political system is highly fractionalized by numerous political parties that range from far-right conservatism to left-wing radicalism.

Education, health, and welfare. Primary education from six to 14 years of age is free and officially compulsory, although the latter is difficult to enforce in some areas. Secondary education, lasting up to four years, is not compulsory. In the late 20th century about four-fifths of the primary-age children were attending school, but only about one-fourth of those of secondary age did so. Most education is state-supported, but private institutions are permitted. Roman Catholic, Protestant, and Jewish organizations are active in primary and secondary education. Adult literacy rates have improved dramatically since the 1950s. In the 1960s about three-fourths of the population was illiterate or semiliterate, but by the 1980s the figure had dropped to about one-third. Higher education takes place in the nation's eight state universities (located in each of the departmental capitals except Pando). There are also technical universities and a Roman Catholic university. The largest university is the University of San Andrés, located in La Paz. The university's library and the National Library of Congress house significant collections.

Health services

There are three kinds of health services—those supported by the state through the Ministry of Public Health and Social Security, those provided by the social security system for its affiliates, and private clinics. In general, the medical services and the hospitals in the cities are adequate. This is not so in rural areas, however, where doctors and nurses are in short supply. Traveling health workers provide primary health care and give advice on hygiene and nutrition in colonization zones. Some major problems persist, however. Communicable diseases such as malaria and the deadly Chagas' disease are still prevalent in the Oriente, and respiratory diseases are widespread. Malnutrition is a problem in some areas.

CULTURAL LIFE

Traditional culture. Bolivian culture consists of a mixture of Indian culture with the Mediterranean culture brought by the Spaniards. On religious feast days, for example, pagan pre-Columbian rites are still practiced, and the Indians express themselves through dances and songs that blend the two cultures. In such festivities, some symbolic dress presents the Indian interpretation of European attitudes: the dance of the *palla-palla* or *loco palla-palla* caricatures the European invaders, the dance of the *waka-tokoris* satirizes bullfights, and the *morenada* mocks white men, who are represented leading imported African slaves. Indian musical instruments are used to accompany these dances. In the music itself, the mixture of cultures is also evident, since many of the tunes are based on Spanish dances. The more commonly used instruments are the sicu or zampoña (panpipes) and the *kena, tarka,* and *pinkillo* (vertical flutes). Percussion instruments of various sizes are used, including skin drums, bronze gongs, and copper bells. The costumes, highly embroidered and colourful, imitate the dress of the pre-Columbian Indian or the dress of 16th-century Spaniards. Spain's influence is clearly seen in the charango, a stringed instrument peculiar to Bolivia. It has five double strings and is similar in shape to the guitar, although much smaller. Its sound box is made from the shell of an armadillo.

Traditional dress

The daily dress of highland Indian women in both the urban and rural regions remains traditional: very full skirts (polleras) and colourful shawls. The latter are usually stuffed with goods being taken to market for sale, as well as with fresh purchases, extra clothing, and a baby, all in a carefully balanced bundle on the back, leaving both hands free. Hats always complete the outfit, their shapes varying with the different regions of Bolivia.

Since the 1940s, Indian culture has blossomed. Previously, Indians had attempted to imitate Europeans, in custom as well as in dress. By the early 1970s, however, Indian values had been reestablished; Indian music rose to a higher standard, painters abandoned the imitation of European fashions, and some of the characteristics of Indian culture reemerged in the general life-style. The Academy of the Aymara Language in La Paz is dedicated to the preservation of the purity of the Aymara language.

The arts. The National Academy of Fine Arts in La Paz offers courses in music, painting, sculpture, and ceramics. There are two galleries of pictorial arts and permanent collections in the National Museum of Art and the National Museum of Archaeology (Museo Tiahuanaco). Jewelry in silver and gold, with pre-Columbian decorations and styles, has been made since colonial times. The local markets offer a profusion of colourful handicrafts and fine wood carvings.

In the city of Potosí the impressive Royal Mint (Casa Real de Moneda), founded in the 16th century and rebuilt in the 18th century, has been restored, and its great halls are used to house a collection of fine colonial paintings. Various old temples have also been restored. In La Paz is the National Symphony Orchestra, and at the University of San Andrés a choral group specializes in Indian musical themes.

Recreation. Folk festivals are held periodically, the most important one of which is held in Oruro during the carnival holidays. Many Indian music and dance groups compete, providing a magnificent display of costumes and decorations. Football (soccer) is the national sport, and Bolivian national teams compete in international matches. A number of other spectator sports common to the West may be seen in the large cities. There are occasional bullfights.

Press and broadcasting. Bolivia has an active and constitutionally free press, which has been subject to periodic censorship during dictatorial regimes. Each of the departmental capitals (except Cobija, Pando) has at least one daily newspaper. *El Diario* (the oldest), *Presencia* (Roman Catholic), *Hoy,* and *Ultima Hora,* all in La Paz, have the largest national circulations. Sucre's daily, *Correo del Sud,* was established only late in the 20th century. Radio stations, many of them commercial, are widespread in Bolivia, broadcasting in Spanish, Aymara, and Quechua. The state-owned National Television Company (Empresa Nacional de Televisión), covering national and world events, sports, the arts, documentaries, and general entertainment, operates stations in most of the major cities. The University Television Service (Televisión Universitaria) also presents educational programs. By the 1980s, however, most Bolivians still did not own a television set. The percentage of viewers in the urban areas is, of course, much higher than in rural districts, with television being found not only in homes but in public and commercial establishments as well.

Newspapers

For statistical data on the land and people of Bolivia, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL. (J.V.F./V.A.U.)

History

EARLY PERIOD

Pre-Columbian civilizations

Bolivian society traces its origins to the advanced pre-Columbian civilizations of South America. The high Bolivian plateau known as the Altiplano was already densely populated several centuries before the Spanish conquest in the 16th century.

From the 7th century the Tiahuanaco Empire, the first of the great Andean empires to extend over both the Peruvian coast and highlands, had its centre in the Altiplano region. By the 11th century it had reached its apogee and was replaced by simpler regional states.

In the centuries that followed the collapse of Tiahuanaco, the Bolivian highland region maintained its dense populations and technological civilizations with irrigation agriculture. By the 15th century the region was mostly controlled by some 12 nations of Aymara-speaking Indians. Competitors to the Quechua-speaking nation at Cuzco, in what is now Peru, these Aymara tribes fought with the latter for control of the central Andean highland region. Though the Aymara nations were eventually dominated by Cuzco, they nevertheless remained the most important non-Quechuan group within the expanding Inca Empire; because of their importance they were the only conquered coastal or highland peoples able to retain their language and cultural identity to such an extent that their Aymara language survived the Spanish conquest. But the Aymara were forced to accept a large body of Quechua-speaking immigrants within their midst as part of a deliberate Incan policy of colonization. It was this early pattern of colonization and nonassimilation that gave Bolivia its current linguistic and cultural identity (the two major Indian languages in Bolivia today are Quechua and Aymara).

Part of the Spanish Empire

The southern Andean valleys and central plateau of Bolivia, with their dense Indian populations, became a core population area within the Spanish Empire after the conquest. To the wealth represented by Indian labour there was added the mineral wealth from newly discovered (1545) silver deposits of Potosí—the largest silver mines then known in the Western world. The arid, high-altitude mines of Potosí, along with others discovered near the town of Oruro (founded in 1606), were supplied with food and other basic necessities by such supportive towns as Chuquisaca (founded in 1538), La Paz (1548), and Cochabamba (1571). From the 16th to well into the 18th centuries, this southern Andean area, known then as Charcas, or Upper Peru, was one of the wealthiest and most densely populated centres of Spain's American empire. Its mines were supplied with forced drafts (mitas) of Indian labourers from throughout the Andes; and by the middle of the 17th century its central mining city of Potosí was the largest city in the Americas, with a population estimated at 150,000. By the last quarter of the 18th century this famous mining zone was in decay. The exhaustion of the richest and most accessible veins, the poverty of the miners and their ignorance of advanced technology, and the absorption of colonial capital into the thriving international commerce of the late 18th century all contributed to its decline. Although Potosí continued to be Upper Peru's most important economic centre, the intellectual and political focus of the area was Chuquisaca (also known in the colonial period as Charcas and La Plata and, since independence, as Sucre). With its academies and university, Chuquisaca was the major educational centre for the entire Río de la Plata region; it also served as the seat of Upper Peru's government, which was known from its foundation in 1559 as the Audiencia of Charcas. The Audiencia was first placed under the Viceroyalty of Peru at Lima, but in 1776 it was finally shifted to a new viceroyalty established at Buenos Aires.

In the late 1770s and early 1780s abuses by petty officials instigated Indian uprisings that were carried out in the hope of reestablishing the old Inca Empire. Severe disorders were widespread in the highlands and produced many casualties; La Paz was besieged twice for several months. Eventually the Indian leaders were defeated and executed.

Independence movement

In 1809 Chuquisaca and La Paz became two of the earliest cities in Spanish America to rebel against the Spanish authorities appointed by the new Napoleonic ruler of Spain. Authorities often considered this action to be the beginning of the Wars of Independence in Latin America. Although viceregal authorities in Lima were able to put down the rebellions, similar uprisings were successful in the viceregal capital of Buenos Aires. From that city several revolutionary armies were dispatched without success to liberate Upper Peru. However, the guerrilla armies that formed in the rugged countryside after the events of 1809 kept the revolutionary movements alive in Upper Peru for some 16 years. In 1825 the Bolivian army from the north, under the leadership of Marshal Antonio José de Sucre, liberated Upper Peru. The military victory was facilitated by the continual defections of Upper Peruvian royalists. Mostly a native elite, the defectors convinced Simón Bolívar and Sucre to allow Upper Peru to become independent rather than be joined to either Peru or Argentina. On Aug. 6, 1825, an Upper Peruvian congress declared Bolivia independent. Few of the guerrilla commanders, representing a more humble constituency, were able to become part of the new order, controlled by the elite.

BOLIVIA FROM 1825 TO C. 1930

In recognition of Bolívar's support, the leaders of the congress named the new republic Bolivia after the liberator and invited Sucre, his chief aide, to be the first president.

Foundation and early national period. But the new republic was not as viable as its leaders fervently hoped it would be. It was economically retarded, despite the legendary colonial wealth and prominence of the region. The late 18th-century mining decline had given way to severe depression as a result of the Wars of Independence. Between 1803 and 1825 silver production at Potosí declined by more than 80 percent; and by the time of the first national census in 1846, the republic listed more than 10,000 abandoned mines.

Incapable of exporting its silver at the levels of colonial production, Bolivia rapidly lost its previously advanced economic standing within Spanish America. Already by the end of the colonial period, such marginal areas as the Río de la Plata and Chile were forging ahead on the basis of meat and cereal production. Bolivia, on the other hand, was a net importer of basic foods, even those exclusively consumed by its Indian population. And none of its mineral resources was of sufficient value to overcome the high transportation costs involved in getting the minerals to the coast. The Bolivian republic, with little trade to tax and with few resources to export beyond its very modest precious metals production, was forced to rely on direct taxation of its Indian peasant masses, who made up more than two-thirds of the estimated 1,100,000 population in 1825. Until well into the last quarter of the 19th century, this regressive Indian taxation was a large source of national government revenue. With the more progressive South American states relying almost exclusively on import and export taxes of a constantly expanding international commerce, the Bolivian state rapidly lost its prominent position within the continent and became known as one of the most backward of the new republics.

Economic decline and political stagnation

This economic decline was mirrored by political stagnation. Bolivia emerged initially with a series of military strongmen (caudillos), among whom was Marshal Andrés de Santa Cruz, president from 1829 to 1839. Temporarily reorganizing the war-torn Bolivian economy and state finances, Santa Cruz in the 1830s was able to unite Bolivia with Peru, through the successful overthrow of the Lima regime of General Agustín Gamarra, into a government known as the Confederation (1836–39). But Chilean military intervention destroyed the Confederation attempt; Bolivia quickly turned in upon itself and thenceforth abandoned all attempts at international expansion.

Bolivia's efforts for the next half-century were primarily to integrate its far-flung regions into a coherent relationship between the core of the republic, the Altiplano, and the eastern Andean valleys. This attempt was doomed to failure, however, because Bolivia lacked the population and resources to exploit either its Amazonian or its Pacific frontiers. Despite the enormous wealth in nitrates and guano available on the Bolivian Pacific coast, the nation was incapable of exploiting them even with the help of foreign capital. What little capital was available within Bolivia's upper class was committed to Altiplano mining. It was Peruvians, Chileans, North Americans, and Englishmen who exploited these resources. Between the Chilean War of the Confederation (1838–39) and the outbreak of the War of the Pacific (1879), Chile successfully expanded its claims, both by diplomatic pressure and, finally, by war, against Bolivian sovereignty in the enormous Bolivian Pacific coastal territories.

War with Chile

Loss of the coastal region. The War of the Pacific (1879–84) had its origins in heavy Anglo-Chilean investments in Bolivia, beginning in the 1840s with those in the guano deposits in Bolivia's Atacama coastal province. With the discovery of nitrate deposits in the 1860s, Chilean aggressive expansion along the coast increased even further. By a series of treaties Chile expanded its territorial claims and obtained commercial concessions within Bolivian territory. As a response to this pressure, Bolivia signed an alliance treaty with Peru in 1873; but this did not intimidate the Chileans. When the Bolivian government attempted to increase taxation on Chilean nitrate companies working in Bolivia, Chile unilaterally invaded Bolivian territory in 1879 and forced a war with both Bolivia and Peru. In May 1880, at the Battle of Tacna, Chile defeated a combined Bolivian-Peruvian army, ending effective Bolivian resistance. Rather than attack the Andean core of Bolivia, the Chileans ignored Bolivia for the rest of the war and proceeded on a massive invasion of Peru, which resulted in their eventual capture of Lima.

The fall of the Pacific littoral to Chile may, in many ways, have been a blessing for Bolivia. The War of the Pacific marked a major turning point in national history. From the fall of the Confederation to the War of the Pacific, Bolivia had gone through one of the worst periods of caudillo rule in all Latin America during the 19th century. The decades of the 1860s and '70s, however, were those in which Andean silver mining had revived under the impact of new capital inputs from Chile and Great Britain. By the time of the War of the Pacific, international market conditions for silver and the introduction of new technology and capital had greatly revived the national mining industry. The War of the Pacific enabled the new mining entrepreneurs to capture political control of the nation and to break the hold of the by-then discredited barracks officers on national political life.

Era of civilian government

Formation of Liberal and Conservative parties. Starting in 1880, under the presidency (1880–84) of Narciso Campero, Bolivia moved into an era of civilian government with the national upper class dividing into Liberal and Conservative parties, which then proceeded to share power. This intraclass political party system finally brought Bolivia the stability it needed for economic development. Though the parties split on personality and anticlerical issues, they were identical in their desire to promote economic growth. From 1880 to 1899 the nation was ruled by the Conservatives, whose principal function was to encourage the mining industry through the development of an international rail network.

It was thus an economically expanding nation that the Liberals inherited when they seized power from the Conservatives in the so-called Federal Revolution of 1899. This revolt, although supposedly fought over the permanent placing of national institutions in the city of Sucre or La Paz, was primarily a power struggle between the Conservative and Liberal parties. Unfortunately for the Conservatives, their strength was too closely tied to the traditional Chuquisaca elite, much of which was coterminous with the silver-mining class. The Liberals, however, had the bulk of their strength in La Paz, which by this period was three times the size of Sucre and had the largest urban concentration (some 72,000 persons) in the national population, estimated at 1,700,000 persons in 1900.

Increase in tin mining. The Liberal victory was also closely associated with a basic shift in the mining economy. As the world silver market began to collapse in the 1880s and early 1890s, a major shift to tin mining began on the Bolivian Altiplano. Found in association with silver, tin did not become an important product until the end of the 19th century, when demand suddenly soared in all the major industrialized countries. Thus by 1900 tin had completely superseded silver as Bolivia's primary export, accounting for more than 50 percent of national exports.

The tin-mining elite

The shift to tin mining not only occurred at the same time as the Liberal revolt and was closely associated with the new party but it also brought about a basic change within the capitalist class in Bolivia. Whereas the silver-mining elite had been almost exclusively Bolivian, the new tin miners were far more cosmopolitan, including, in the early years, foreigners of all nationalities as well as some new Bolivian entrepreneurs. Tin mining itself absorbed far more capital and produced far more wealth than had the old silver-mining industry, and the new companies that emerged became complex international ventures directed by professional managers.

Given this new economic complexity and the political stability already achieved by the Conservatives and perpetuated by the Liberals, the tin-mining elite found it profitable to withdraw from direct involvement in national political life. Whereas Bolivian presidents under Conservative rule in the 19th century had either been silver magnates themselves (Gregorio Pacheco, 1884–88; Aniceto Arce, 1888–92) or closely associated with such magnates as partners or representatives (Mariano Baptista, 1892–96; Severo Fernández Alonso, 1896–99), the Liberals and subsequent 20th-century presidents were largely outside the mining elite. No tin magnate actively participated in leadership positions within the political system. Rather, they came to rely on a more effective system of pressure group politics.

Liberal rule, 1899–1920. The primary task of the Liberal politicians, who ruled Bolivia from 1899 to 1920 under the leadership of Ismael Montes (twice president: in 1904–08 and 1913–17), was to settle Bolivia's chronic border problems and to continue and expand the communications network initiated by the Conservatives. A definitive peace treaty was signed with Chile (1904), accepting the loss of all Bolivia's former coastal territories. Also, the Acre problem was resolved: this involved an unsuccessful attempt by the central government to crush an autonomist rebellion (1889–1903) in the rubber-boom territory of Acre on the Brazilian border. Brazil's covert support of the rebels and the defeat of Bolivian forces finally convinced the Liberals to sell the territory to Brazil in the Treaty of Petrópolis (1903). As a result of the financial indemnities provided by both treaties, Bolivia was able to finance a great railroad construction era—by 1920 most of the major cities were linked by rail, and La Paz was connected to the two Chilean Pacific ports of Antofagasta and Arica; new lines had been begun or completed to Lake Titicaca and thus the Peruvian border, and to Tarija and the Argentine frontier.

Railroad-building era

The period of Liberal rule was also the most calm in Bolivian political history. Dominated by the figure of Montes, the Liberal Party easily destroyed the federalists, who had supported the party during the revolution, and Liberal success led to the total collapse of the Conservative Party. Not until 1914 was an effective two-party system again established, when many of the political "outs," along with a large number of new and younger elements, finally organized the Republican Party. Like its predecessors, the Republican Party was a white, upper- and middle-class grouping, with a fundamental belief in liberal and positivist ideologies.

The Republican Party. The abrasive quality of the strong-willed Montes and the disintegration of the ruling Liberal Party finally permitted the Republicans to stage a successful coup d'état, in 1920, and become the ruling party. Upon achieving political power, however, the new party immediately split into two warring sections based on a conflict of personalities and led by two Montes-style politicians—Juan Bautista Saavedra, a La Paz lawyer who captured control of the Republican Party and was national president from 1930 to 1934, and Daniel Salamanca, a Cochabamba landowner who took his following into a separate party, the so-called Genuine Republican Party (Partido Republicano Genuino). The rivalry between these two men became the dominant theme in Bolivian politics for the next decade, until the Salamanca forces captured the presidency.

Below the surface of this political battle of personalities, the national economy in the 1920s was undergoing serious change. The early years of the decade had witnessed a brilliant post-World War I recovery of the Bolivian tin-mining industry, which led to the achievement by 1929 of its highest production figures. This enormous output occurred, however, in a period of steady price decline (a trend that continued long after the Great Depression of the 1930s). By 1930 the international tin market was in a serious crisis, and Bolivian production suffered. The year 1930 marked the end of major new capital investment in tin mining; thereafter the industry in Bolivia would become an ever higher-cost producer of lower-grade ores for the international market.

BOLIVIA SINCE 1930

The installation of Salamanca in the presidency after the revolt of 1930 seemingly involved little change in traditional Bolivian government. But the Great Depression cut brutally into national income and forced the closing of a large part of the vital mining industry. Salamanca was thus forced to take new measures. Attempting a policy of inflation and money manipulation, he ran into bitter hostility from the Liberal Party, his key partner in the 1930 overthrow of the regular Republican Party. The opposition of these two forces in the central government led to a tense political climate, the result of which was to force Salamanca to accept a Liberal veto over internal economic decisions. He refused, however, to permit the Liberals to join his cabinet; rather, he sought to overcome Liberal congressional control and to destroy growing strike movements by turning national attention to other themes. Salamanca had the traditional recourse to patriotism and foreign war open to him in a long-standing border conflict with Paraguay.

Conflict with Paraguay

Already in the mid-1920s, Bolivia and Paraguay both had begun a major program of fort construction in the largely uninhabited and poorly demarcated Chaco Boreal territory, on the southeastern Bolivian frontier. At the height of the Depression, Salamanca had advocated an even heavier armament and fortification program and had contracted for major European loans. In June 1932 a border incident developed between the two states, and Salamanca deliberately provoked a full-scale Bolivian reprisal, which inevitably led to open war between the two nations.

The Chaco War. The Chaco War was a long and costly disaster for Bolivia. In the three years of bitter fighting on its southeastern frontiers, Bolivia lost 100,000 men dead, wounded, deserted, or captured; it also lost far more territory than Paraguay had claimed even in its most extreme prewar demands. The fact that Bolivia had entered the war with a better equipped and supposedly far better trained army only aggravated the sense of frustration among the younger literate veterans—the so-called Chaco generation—at the total failure of Bolivian arms. Charging that the traditional politicians and the international oil companies had led Bolivia into its disastrous war, the returning veterans set up rival Socialist and radical parties and challenged the traditional political system.

Return to military rule. The initial result of this challenge was the overthrow of the civilian government and the first advent of military rule in Bolivia since 1880. In 1936 the younger army officers seized the government, and under two leaders, Colonel David Toro in 1936–37 and Major Germán Busch in 1937–39, they tried to reform Bolivian society. Little was accomplished during this so-called era of military socialism except for the confiscation of the Standard Oil Company holdings, the creation of an important labour code, and the writing in 1938 of an advanced, socially oriented constitution.

Political chaos

Rise of new political groups. Beginning in the 1940s, civilian dissident groups finally began to organize themselves into powerful national opposition parties. The two most important of these were the middle-class and initially fascist-oriented Nationalist Revolutionary Movement (Movimiento Nacionalista Revolucionario; MNR) and the Marxist and largely pro-Soviet Party of the Revolutionary Left (Partido de la Izquierda Revolucionaria; PIR). Both groups established important factions in the national congress of 1940–44. In 1943 the civilian president General Enrique Peñaranda was overthrown by a secret military group, Reason for the Fatherland (Razón de Patria; Radepa), which allied itself with the MNR and tried to create a new-style government. Under Colonel Gualberto Villaroel (1943–46) little was accomplished, except for an initial political mobilization of the Indian peasants by the MNR. Opposed as fascist-oriented by the right and left, the Villaroel government was overthrown in 1946, in a bloody revolution in which Villaroel was hanged before the presidential palace by revolutionary crowds.

During the next six years, the PIR tried to rule in alliance with many of the older parties but failed. It was eventually dissolved and replaced in early 1950 by the more radical Bolivian Communist Party. The Conservative parties proved unable to control the situation; and, after the MNR won a plurality victory in the presidential elections of 1951, the military intervened directly and formed a military junta government. As for the MNR, its disaster under Villaroel led it to disassociate itself from its fascist wing and to seek an alliance with a small Trotskyite party, which had important mine-union support. The resulting alliance brought the labour leader Juan Lechín into the MNR. After several unsuccessful revolts, each more violent than the preceding one, the MNR, in several days of fighting, finally overthrew the military regime in April 1952. During this struggle, workers, civilians, and peasants were armed and the army was almost totally destroyed.

The Bolivian National Revolution (1952). Thus in April 1952 began the so-called Bolivian National Revolution, which became one of Latin America's most important social revolutions. The MNR and its mine-worker and peasant supporters were pledged to a fundamental attack on the tin-mining industry and its allied political supporters. In October 1952 the three largest tin-mine companies were nationalized. In August 1953 came one of the most far-reaching land-reform decrees ever enacted in the Western Hemisphere. Universal suffrage was granted with the abolition of literacy requirements. Not only were Indians granted land, freed from servile labour obligations, and given the vote, but they were also given large supplies of arms. From that point on, the Indian peasants of Bolivia became a powerful, if largely passive, political force upon which all subsequent governments based their strength.

Social reforms

Post-1952 regimes. The most important leader of the MNR, Víctor Paz Estenssoro, was president of Bolivia in 1952–56 and instituted the most revolutionary part of the party's program. In 1956 he was replaced by the more conservative Hernando Siles Zuazo, whose primary concern was to stop an inflation that had followed the 1952 revolution. That inflation had in many ways completed the revolutionary process by virtually destroying the older middle-class supporters of the MNR. With massive financial support from the United States, the inflation was stopped; but most of the advanced social programs of the revolution were suspended as well. The government ended worker co-administration of the nationalized mine companies and cut back on expanded social services. It also invited U.S. petroleum companies back into Bolivia for the first time since 1937, when Standard Oil of Bolivia had been confiscated by the Toro government.

With the return of Paz Estenssoro to the presidency in 1960, the consolidation achieved by Siles Zuazo was further developed, with the power of the army being revived under U.S. support. The attempt of Paz Estenssoro to renew his presidential term for another four years, in 1964, led to the splintering and temporary destruction of the MNR and to the overthrow of his government by the military.

The Barrientos regime

With the support of many conservative elements and the peasant masses, the vice president, General René Barrientos, seized the government and proceeded to dissolve most of the organized labour opposition, marking the beginning of a reign of military leaders. From 1964 until his death in 1969, Barrientos continued with the process of conservative economic reform and political retrenchment and attempted to demobilize all popular groups except the peasants, who had risen to some power as a result of the National Revolution. A poorly organized antigovernment campaign, led by the Argentinian-born revolutionary Ernesto (Che) Guevara, was destroyed in 1967, largely because of its failure to mobilize the peasants.

The death of Barrientos brought the vice president, Luis Adolfo Siles Salinas, into office; he was forcibly replaced in mid-1969 by General Alfredo Ovando Candía, who in turn was forced out of office, after he had nationalized Gulf Oil Company holdings, by the more radical General Juan José Torres in October 1970. Of the several military regimes that governed between 1964 and 1979, that led by Torres was the most radical; for a time the Torres government replaced Congress with a workers' soviet. In 1971 Torres was replaced by Colonel Hugo Banzer Suárez, and the most repressive regime of the period came to power. During the next seven years the labour movement was suppressed, the mines were occupied by troops, all civil rights were suspended, and the peasant syndicates were prohibited. Nevertheless, this was also an era of unprecedented growth in the Bolivian economy, fueled by the sudden increase in world mineral prices and the completion of some of the basic changes in the social and economic infrastructure begun with the National Revolution of 1952. These long-term economic changes involved the relative decline of the importance of tin and the emergence of commercial agricultural exports for the first time in Bolivian republican history. It was also a period of extraordinarily rapid increase in the national population, which between 1950 and 1976 achieved an annual net growth rate of 2.1 percent. Finally, the Banzer regime was unique in contemporary Bolivian affairs because it provided the new commercial agricultural interests of the Santa Cruz region with national representation.

In 1978 the old MNR reemerged, and a complex set of new political parties and movements developed. In the national elections of 1978 and 1979, these new groups gained wide support, and the electorate showed an even balance between conservative and radical positions. Moreover, peasants for the first time no longer voted in block but were as equally divided as the urban populace.

These democratic changes were fragile, however. The results of the elections of July 9, 1978, were voided in the wake of charges of fraud. Banzer resigned later that month under threat of a coup. The unstable state of government went on unabated over the next several years as the country's leadership continued to change hands rapidly. In November 1979 Lidia Gueiler Tejada was chosen by military, political, and union leaders to serve as interim president, the first woman to hold the country's highest office. She served until elections were held in June of the following year. None of the presidential candidates in those elections received a majority vote, however, and, instead of an orderly return to democratic government, the army staged another coup, established a military junta, and declared a state of siege. The new regime immediately lost credibility in the international community because of its repressive measures and because U.S. officials implicated some of its leaders in illegal cocaine trafficking.

Upheaval under military governments

The cycle of coups and overthrows continued. Economic conditions worsened, fueled by the mismanagement of successive military regimes. In September 1982 the military reinstated the Congress that had been elected in 1980. They installed as president Siles Zuazo, who had been the victor in the 1980 elections. Siles Zuazo managed to stay in office until the next elections, held in July 1985, despite massive inflation, crippling strikes to protest his strict economic controls, and disastrous crop failures caused by El Niño, the warm ocean current that sometimes brings climatic upheavals to western South America. As the new elections approached, Bolivia was a country in chaos. Its devastating inflation rate, the world's highest, had gone into quintuple digits; its major source of foreign exchange was derived from cocaine dealing, which the government was reluctant to curtail because of the hardships it would impose on the farmers who grew the coca plant; and strikes and violent demonstrations were frequent occurrences.

The 1985 elections once again brought Paz Estenssoro back to the presidency; it was the first orderly transfer of power in some 25 years. He was chosen by the Congress to lead despite the fact that another candidate, former president Banzer, had a larger plurality. Paz Estenssoro, who once had brought liberal reforms to the country, exacted extreme austerity measures that were greeted with renewed protests and strikes. The new president, however, rather than bending to public pressure, held to his plan, and after several months Bolivia witnessed a turn in the economy, characterized chiefly by a reduction of its ruinous hyperinflation rate. Although this was an encouraging step, Bolivia still faced numerous debilitating problems, particularly that of unemployment, which had been exacerbated by the collapse of world tin prices, causing tens of thousands of miners to lose their jobs. The difficult fight against the large underground cocaine economy was somewhat strengthened, partly as a result of pressure from the United States. Actions taken included the eradication of illegal crops and processing laboratories, done with the aid of U.S. troops and helicopters, and implementation of a program of crop diversification.

Crackdown on illegal cocaine activity

The presidential elections of May 1989 were peaceful, but, because none of the three candidates received a majority as required by the constitution, the choice fell to Congress. Aided by the support of the second-place finisher and former president, Banzer Suárez, third-place finisher Jaime Paz Zamora of the Movement of the Revolutionary Left (Movimiento de la Izquierda Revolucionaria) became the new president on August 6. In most respects, Paz Zamora's policies were a continuation of his predecessor's (a distant relative), except for a new emphasis on national production of food and raw materials. Under the threat of the loss of U.S. aid, Paz Zamora vowed that Bolivia would continue to help fight the drug trade and seek alternative livelihoods for its thousands of coca farmers.

(C.W.A./Ed.)

For later developments in the history of Bolivia, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 951, 964, 966, and 974, and the *Index.*

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Bombay

Bombay (Mumbai in the Marāṭhī language), the capital of Mahārāshtra state, on the west coast of India, is the country's financial and commercial centre and the principal port on the Arabian Sea. It is one of the largest and most densely populated cities in the world. Bombay is located on a site of ancient settlement and took its name from the local goddess Mumba—a form of Pārvatī, the consort of Śiva (Shiva), one of the principal gods of Hinduism—whose temple once stood in what is now the southeastern section of the city.

Bombay has long been the centre of India's cotton-textile industry, but its other manufacturing industries are now well-diversified and its commercial and financial institutions strong and vigorous. It suffers, however, from the chronic ills of most large, expanding industrial cities—air and water pollution, slums, and overcrowding. Expansion of Bombay is confined by its island location, and the city, which has one of the highest population densities in the world, seems ready to burst at its seams.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

The city site. The city of Bombay occupies a peninsular site originally composed of seven islets lying off the Konkan Coast of western India. Since the 17th century, drainage and reclamation projects, as well as the construction of causeways and breakwaters, have joined the islets to form a larger body known as Bombay Island. East of the island are the sheltered waters of Bombay Harbour. Bombay Island consists of a low-lying plain, one-fourth of which lies below sea level; the plain is flanked on the east and west by two parallel ridges of low hills. Colāba Point, the headland formed on the extreme south by the longer of these ridges, protects Bombay Harbour from the open sea. The western ridge terminates at Malabār Hill, 180 feet (55 metres) above sea level, which is one of the highest points in Bombay. Between the two ridges lies the shallow expanse of Back Bay. On a slightly raised strip of land between the head of Back Bay and the harbour is the fort, the original nucleus around which the city grew; it is now chiefly occupied by public and commercial offices. From Back Bay, land slopes northward to the central plain. The extreme north of Bombay Island is occupied by a large salt marsh.

The old city covered about 26 square miles from Colāba in the south to Māhīm and Sion in the north. In 1950 Bombay expanded northward with the inclusion of the large island of Salsette, which was joined to Bombay Island by a causeway. By 1957 a number of suburban municipal boroughs and some neighbouring villages were incorporated into Greater Bombay. The Bombay metropolitan region has continued to expand, and by the mid-1980s the Bombay metropolitan region comprised an area of 233 square miles (603 square kilometres). During the early 1970s, in an effort to relieve congestion, Salsette Island was linked with the mainland by a bridge across Thāna Creek, the headwaters of Bombay Harbour.

Bombay Harbour

The natural beauty of Bombay is unsurpassed by that of any city in the region. The entrance into Bombay Harbour from the sea discloses a magnificent panorama framed by the Western Ghāts (mountains) on the mainland. The wide harbour, studded with islands and dotted with the white sails of innumerable small craft, affords secure shelter to ships, particularly when storms lash the coast. The largest of the harbour's islands is Elephanta, which is famous for its 8th- and 9th-century cave temples.

Typical trees in the city include coconut palms, mango trees, and tamarinds, as well as banyan trees. Salsette Island was once the haunt of wild animals, such as tigers, leopards, jackals, and deer, but these are no longer found on the island. Animal life now consists of cows, oxen, sheep, goats, and other domestic species. Birdlife includes vultures, pigeons, cranes, and ducks.

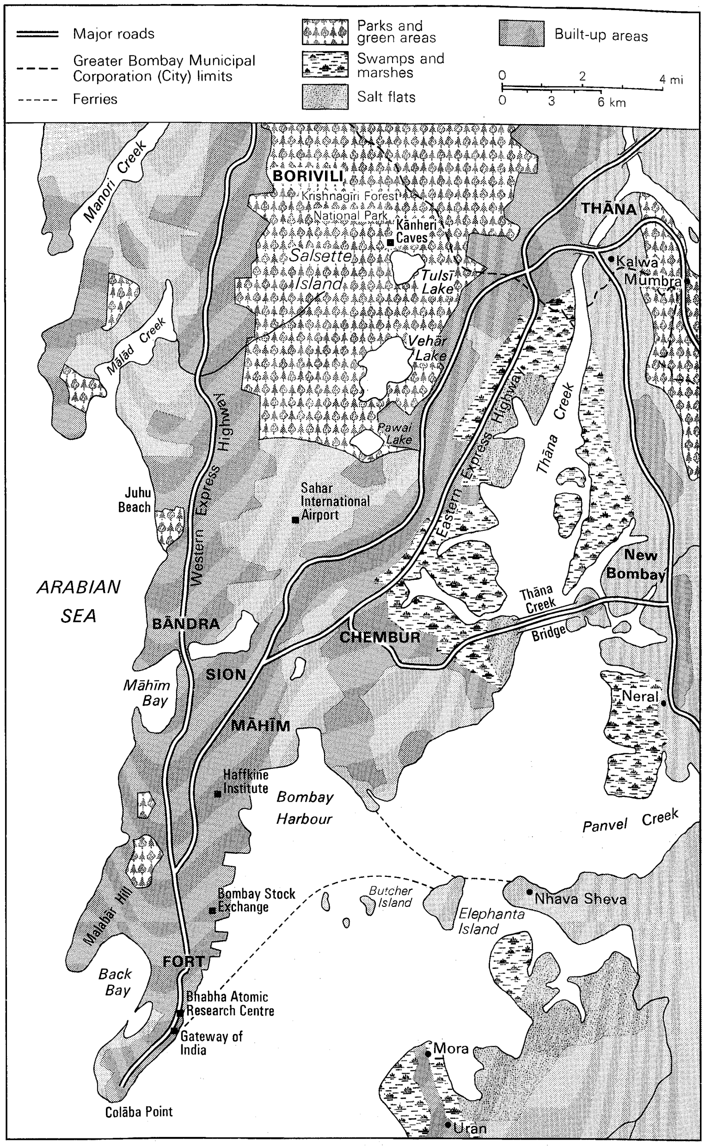
Climate. The climate of Bombay is hot and humid. There are four seasons. Cold weather prevails from December to February, and hot weather from March to May. The rainy season, brought by monsoon winds from the southwest, lasts from June to September and is followed by the post-monsoon season, lasting through October and November, when the weather is again hot. Mean monthly temperatures vary from 91° F (33° C) in May to 67° F (19° C) in January. Annual rainfall is 71 inches (1,800 millimetres), with an average of 24 inches falling in July alone.

The city layout. The older part of Bombay is much built up, but more affluent areas, such as Malabār Hill, contain some greenery; there are also a number of open playgrounds and parks. Bombay's history of burgeoning urbanization has created slums in sections of the city. An alarming rate of air and water pollution has been caused by the many factories still crowding the city, the growing volume of motor-vehicle traffic, and the nearby oil refineries.

The financial district is located in the southern part of the city (around old Fort Bombay). Farther south (around Colāba) and to the west along Netaji Subhas Road (Marine Drive) and on Malabār Hill are residential neighbourhoods. To the north of the fort area is the principal business district, which gradually merges into a commercial-residential area. Most of the older factories are located in this area. Still farther north are residential areas, and beyond them are recently developed industrial areas as well as some shantytown districts.

Housing

Housing is largely privately owned, though there is some public housing built by the government through publicly funded corporations or by private cooperatives with public funds. But Bombay is very crowded, and housing is scarce for anyone who is not very rich. For this reason, commercial and industrial enterprises find it increasingly difficult to attract middle-level professional, technical, or managerial staff. There is continuous immigration of unskilled labour from the hinterland, and the number of indigent and homeless people is increasing. City planners



Bombay and surrounding area.

have sought to stop this movement and to persuade enterprises to locate across Bombay Harbour in the developing "twin city" of New Bombay by banning the development of new industrial units and the expansion of existing ones inside the city. But the ban has been largely breached in practice whenever entrepreneurs threaten to relocate their businesses to some other part of the country.

Bombay's architecture is a mixture of florid Gothic styles, characteristic of the 18th and 19th centuries, and contemporary designs. The older administrative and commercial buildings are intermingled with skyscrapers and multistoried concrete-block buildings.

THE PEOPLE

Bombay's growth since the 1940s has been steady if not phenomenal. At the turn of the 20th century its population was 850,000; in 1941 it had doubled to 1,695,000; and by 1981 it had grown to more than 8,200,000. The city's birth rate is much lower than that of the nation because of family-planning programs, and the high overall growth rate is largely attributable to the influx of people in search of employment.

Low birth rate

Bombay has one of the highest population densities in the world. In 1981 Greater Bombay had an average of more than 35,000 persons per square mile, and in much of the city's older section at least three times that was recorded, though such areas as Girgaum, Bhendi Bazaar, and Bhuleswar, all near Back Bay, had lower population densities. Some parts of the inner city have nearly one million persons per square mile, perhaps the world's highest density.

The city is truly cosmopolitan, and representatives of almost every religion and region of the world can be found in Bombay. Almost half of the population is Hindu; but the city also encompasses important communities of Muslims, Christians, Buddhists, Jains, Sikhs, Zoroastrians, and Jews.

Almost every Indian language and many foreign languages are spoken. Marāṭhī, the state language, is the dominant Indian language, followed by Gujarati and Hindi. Others include Bengali, Pashto, Arabic, Chinese, English, and Urdu.

THE ECONOMY

Bombay is the economic hub and commercial and financial centre of India. Its economic composition in some respects demonstrates India's peculiar fusion of the nuclear and cow-dung ages. The city contains the Indian Atomic Energy Commission's establishment, which includes nuclear reactors and plutonium separators. In many areas of the city, however, traditional biogenic sources of fuel and energy are still in use.

Industry. The cotton-textile industry, through which the city prospered in the 19th century, is still important but is in relative decline. New growth industries—metals, chemicals, automobiles, electronics, engineering, and a host of ancillary enterprises—as well as urban industries such as food processing, papermaking, printing, and publishing have been at the core of expansion of manufacturing employment.

Commerce and finance. The Reserve Bank of India, the country's Central Bank, a number of other commercial banks, and the Life Insurance Corporation of India—a nationalized enterprise and the country's largest investor in government bonds and private stocks—as well as other major long-term investment financial institutions, are located in Bombay. These, in turn, have attracted a number of financial and business services to the city.

The Bombay Stock Exchange

The Bombay Stock Exchange is the country's leading stock and share market. Though a number of the economic hubs that have sprung up around the country since independence have reduced the exchange's pre-independence stature, it remains the preeminent centre in volume of financial and other business transacted and serves as a barometer of the country's economy.

Transportation. Bombay is connected by a network of roads to the north, east, and south of India. It is the railhead for the Western and Central railways, and trains from the city carry goods and passengers to all parts of India. The Sahar International Airport, on Salsette Island, is an important point of entry for many foreign airlines, and nearby Santa Cruz Airport serves domestic flights. Bombay handles about 60 percent of the international and nearly 40 percent of the domestic air traffic in India. The facilities provided by Bombay Harbour make the city India's major western port. Though other major ports have sprung up on the west coast—Kandla to the north and Goa and Cochin to the south—Bombay still handles more than 40 percent of India's maritime trade.

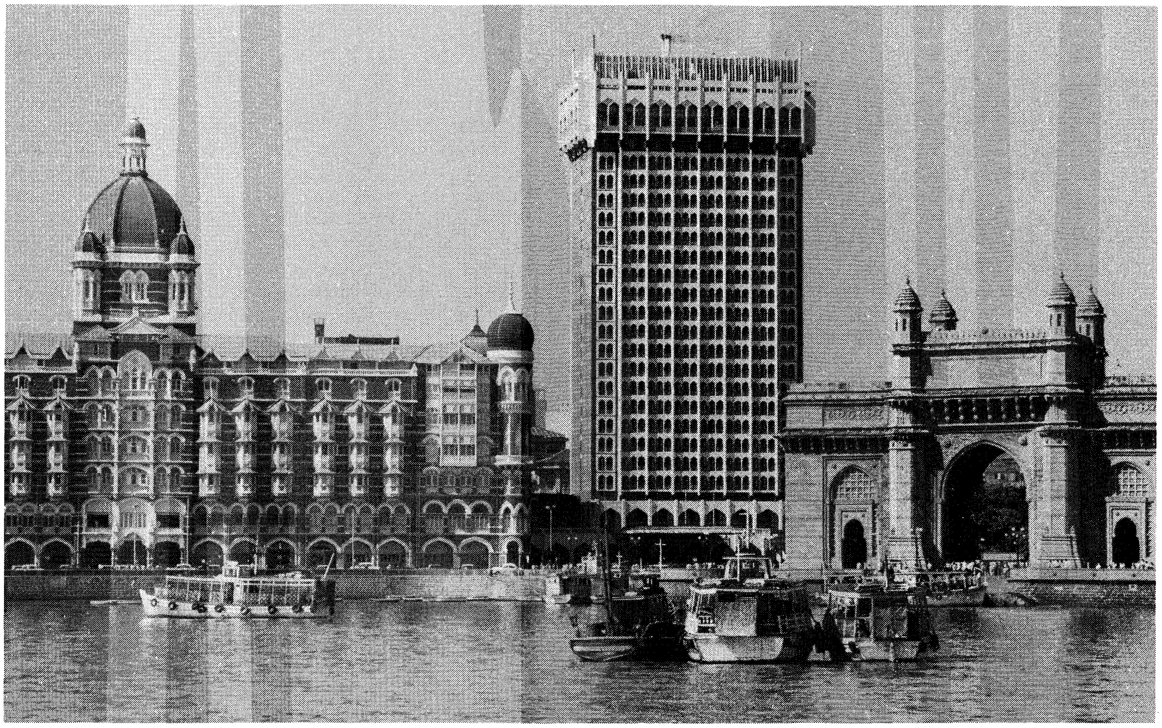
Two suburban electric train systems provide the main public transportation and daily convey hundreds of thousands of commuters from outside the city. There is also a municipally owned bus fleet.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. As the capital of Mahārāshtra state, the city is an integral political division of the state government, the headquarters of which are called the Mantrālaya. The state administers the police force and has administrative control over certain city departments. The central government controls communications such as the post and telegraph system, the railways, the port, and the airport. Bombay is also the headquarters of India's western naval fleet and the base for the Indian flagship.

The Municipal Corporation

The government of the city is vested in the fully autonomous Municipal Corporation of Greater Bombay. Its legislative body is elected on adult franchise every four years and functions through its various standing commit-



Bombay Harbour. The Gateway of India is on the right, and at left and centre are the old and new buildings of the Taj Mahal Hotel.

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tees. The chief executive, who is appointed every three years by the state government, is the municipal commissioner. The mayor is annually elected by the Municipal Corporation; the mayor presides over corporation meetings and enjoys the highest honour in the city but has no real power.

Public utilities. The manifold functions of the corporation include the provision or maintenance of medical services, education, water supply, fire services, garbage disposal, markets, gardens, and engineering projects such as drainage development and the improvement of roads and street lighting.

The Municipal Corporation operates the transport system inside the city and the supply of electricity as public utilities. Electric energy is obtained from a grid system supplied by government and privately owned agencies and is then distributed throughout the city. The water supply, also maintained by the municipality, comes mainly from Tāusa Lake in adjoining Thāna district and from Tulsī and Vehār lakes on Salsette Island. Pawai Lake, originally harnessed for water supply, has proved unsatisfactory because its water is not potable.

Health. The city has more than 100 hospitals, including those run by federal, state, or corporation authorities, and a number of specialized institutions treating tuberculosis, cancer, and heart disease. There are also a number of leading private hospitals. The Haffkine Institute, a leading bacteriologic research centre specializing in tropical diseases, is located there.

Security. The police force is headed by the commissioner of police, who is responsible for law and order in Greater Bombay and is answerable administratively to the home secretary of the state.

Education. Bombay's literacy rate is much higher than that of the nation as a whole. Primary education is free and compulsory; it is the responsibility of the Municipal Corporation. Secondary education is provided by public and private schools supervised by the state government. There are also public and private polytechnic institutes and institutions offering students a variety of degree and diploma courses in mechanical, electrical, and chemical engineering. The Indian Institute of Technology, operated by the central government, is also located in the city. The University of Bombay, established in 1857, has some 130 constituent colleges and more than two dozen teaching departments. Several colleges in Goa are also affiliated with the university.

CULTURAL LIFE

Bombay's cultural life reflects its polyglot population. The city has a number of museums, libraries, literary and other cultural institutions, art galleries, and theatres. Perhaps no other city in India can boast of such a high degree of variety and quality in its cultural and entertainment facilities. Bombay is the stronghold of the Indian film industry and has an open-air theatre. Throughout the year Western and Indian music concerts and festivals and Indian dance shows are performed. The Prince of Wales Museum of Western India, housed in a building of Indo-Saracenic architecture, contains three main sections of art, archaeology, and natural history. Nearby is the Jehangir Art Gallery, Bombay's first permanent art gallery and centre of cultural and educational activities.

Publishing and broadcasting

Bombay is an important centre for the Indian printing industry and has a vigorous press. Newspapers are printed in English, Marāṭhī, Hindi, Gujarati, Sindhi, and Urdu. Several monthlies, biweeklies, and weeklies are also published in the city. The regional station of All-India Radio is centred in Bombay, and television services for the city began in 1972.

Krishnagīri Forest, a national park in the north of Greater Bombay, is a pleasant vacation resort located near the Kānheri Caves, site of an ancient Buddhist university; the more than 100 caves contain gigantic Buddhist sculptures dating from the 2nd to 9th century AD. There are several public gardens, including the Jijāmātā Garden, which houses Bombay's zoo in the city proper; the Baptista Garden, located on a water reservoir in Mazagaon; and the Pherozshah Mehta Gardens, the Kamala Nehru Park, and the Sloping Park—all on Malabār Hill.

Cricket matches, which are popular throughout India, are played at Brabourne and Wankhede stadiums. Athletic and cycling track events are held at the Vallabhbhai Patel Stadium. Juhu Beach is the most fashionable area for bathing and swimming.

History

The Kolis, an aboriginal tribe of fishermen, were the earliest known inhabitants, though Paleolithic stone implements found at Kandivli, in Greater Bombay, indicate human occupation during the Stone Age. The area was known as Heptanesia to the ancient Greek astronomer and geographer Ptolemy and was a centre of maritime trade with Persia and Egypt in 1000 BC. It was part of Aśoka's empire in the 3rd century BC and was ruled in the 6th to 8th century AD by the Cālukyas, who left their mark on Elephanta Island (Ghārāpuri). The Walkeswar Temple at Malabār Point was probably built during the rule of Śilāhāra chiefs from the Konkan Coast (9th–13th century). Under the Yādavas of Devagiri (1187–1318) the settlement of Mahikavati (Māhīm) on Bombay Island was founded in response to raids by the Khaljī dynasty of Hindustān in 1294. Descendants of these settlers are found in contemporary Bombay, and most of the place-names on the island date from this era. In 1348 Bombay was conquered by invading Muslim forces and became part of the kingdom of Gujarāt.

A Portuguese attempt to conquer Māhīm failed in 1507, but in 1534 Sultan Bahādur Shāh, the ruler of Gujarāt, ceded the island to the Portuguese. In 1661 it came under British control as part of the marriage settlement between King Charles II and Catherine of Braganza, sister of the king of Portugal. The crown ceded it to the East India Company in 1668.

Urban expansion

In the beginning, compared to Calcutta and Madras, Bombay was not a great asset to the company but merely helped it keep a toehold on the west coast. On the mainland the Mughals, the Marāṭhās, and the territorial princes in Gujarāt were more powerful. Even British naval power was no match for the Mughals, Marāṭhās, Portuguese, and Dutch. By the turn of the 19th century, external events helped stimulate the growth of the city. The decay of Mughal power in Delhi, the Mughal–Marāṭhā rivalries, and the instability in Gujarāt drove artisans and merchants to the islands for refuge, and Bombay began to grow. With the destruction of Marāṭhā power, trade and communications to the mainland were established and those to Europe were extended; Bombay's prosperity had begun. In 1857 the first spinning and weaving mill was established, and by 1860 Bombay had become the largest cotton market in India. The American Civil War (1861–65) and the resulting cutoff of cotton supplies to Britain caused a great trade boom in Bombay. But, with the end of the Civil War, cotton prices crashed and the bubble burst. By that time, however, the hinterland had been opened, and Bombay had become a strong centre of import trade.

With the opening of the Suez Canal in 1869, Bombay prospered, though slums and unsanitary conditions steadily multiplied with its increasing population. Plague broke out in 1896, and a City Improvement Trust was established to open new localities for settlement and to erect dwellings for the artisan classes. An ambitious scheme for the construction of a seawall to enclose an area of 1,300 acres (525 hectares) was proposed in 1918 but not finished until the completion of what is now Netaji Subhas Road from Nariman Point to Malabār Point—the first two-way highway of its kind in India—after World War II. In the postwar years the development of residential quarters in suburban areas was begun, and the administration of Bombay city through a municipal corporation was extended to the suburbs of Greater Bombay. The city had served as the former capital of Bombay presidency and Bombay state, and it was made the capital of Mahārāshtra state in 1960.

Centre of political activity

During the late 19th and early 20th centuries, Bombay was a centre of both Indian nationalist and regional Marāṭhā political activity. In 1885 the first session of the Indian National Congress (a focus of both pro-Indian and anti-British sentiment until independence) was held in the city, where subsequently, at its 1942 session, the Congress passed the "Quit India" resolution, which demanded complete independence for India. From 1956 until 1960 Bombay was the scene of intense Marāṭhā protests against the two-language (Marāṭhī–Gujarati) makeup of Bombay state, a legacy of British imperialism, which led to the state's partition into the modern states of Gujarāt and Mahārāshtra.

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Boston

Boston, the capital of the Commonwealth of Massachusetts in the northeastern United States, lies on Massachusetts Bay, an arm of the Atlantic Ocean. The city proper—including part of the Charles River, Boston Harbor, and a portion of the Atlantic Ocean—has an unusually small area for a major city, only 46 square miles (119 square kilometres), more than one-fourth of it water.

The area, the people, and the institutions within its political boundaries can only begin to define the essence of Boston. As a city and as a name, Boston is a symbol of much that has gone into the development of the American consciousness, and its presence reaches far beyond its immediate environs. As the spiritual capital of the New England states, as the nation's closest link to its European heritage, as the progenitor of the American Revolution and the nation, and as the earliest centre of American culture, Boston has influenced the country for nearly three centuries. Though Boston, like New England in general, has played a lessening role in national life during much of the 20th century, it remains the focal point of what may be the most diversified and dynamic combination of educational, cultural, and medical and scientific activities in the United States.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

Boston and its sheltered deepwater harbour are ringed by modest hills, from the Middlesex Fells on the north to the Blue Hills on the south.

Original topography

The area of the colonial town. The hilly Shawmut Peninsula, upon which Boston was settled, originally was almost completely surrounded by water. It was connected with mainland Roxbury on the south by a narrow neck of land along the line of present-day Washington Street. To the west of the neck were great reaches of mud flats and salt marshes, which were covered by tides at high water and known collectively as the Back Bay. Beyond this, the Charles River flowed down to the peninsula, further dividing it from the mainland on the north and west as it approached Boston Harbor. On the east the peninsula fronted on the harbour, and Town Cove, jutting in from the harbour, divided Boston into the North End and the South End. The centre of the colonial town was at the present Old State House (1711–47), where the only road from the mainland intersected with the principal approach from the harbour.

Although this original centre and the colonial South End have long been given over to banks, offices, and retail and wholesale trade, a few 18th-century buildings remain: Faneuil Hall (1742–1805), the Old Corner Book Store (1711), the Old South Meeting House (1729), and King's Chapel (1750). The North End is the only part of the early town that has been continuously lived in since the 1630 settlement. Colonial survivals such as the Paul Revere House (*c.* 1680) and Christ Church (1723)—the Old North Church from which lanterns revealed the route of the British march to Lexington in 1775—coexist with the teeming life of a seemingly Mediterranean community.

The considerable extent of shoreline, only a few minutes' walk from any part of the peninsula, provided ample space for wharves and shipyards. From the beginning of the settlement, the shoreline constantly encroached on the harbour as wharves were built and marshy coves were filled. Beyond the original settlement lay Boston Common, a tract of 45 acres (18 hectares) that has remained public open space since its purchase by the town in 1634. Above the Common rose the Trimountain, a three-peaked hill of which Beacon Hill is the only surviving, though greatly reduced, remnant. The other hills became landfill that added to the city's area in the 19th century.

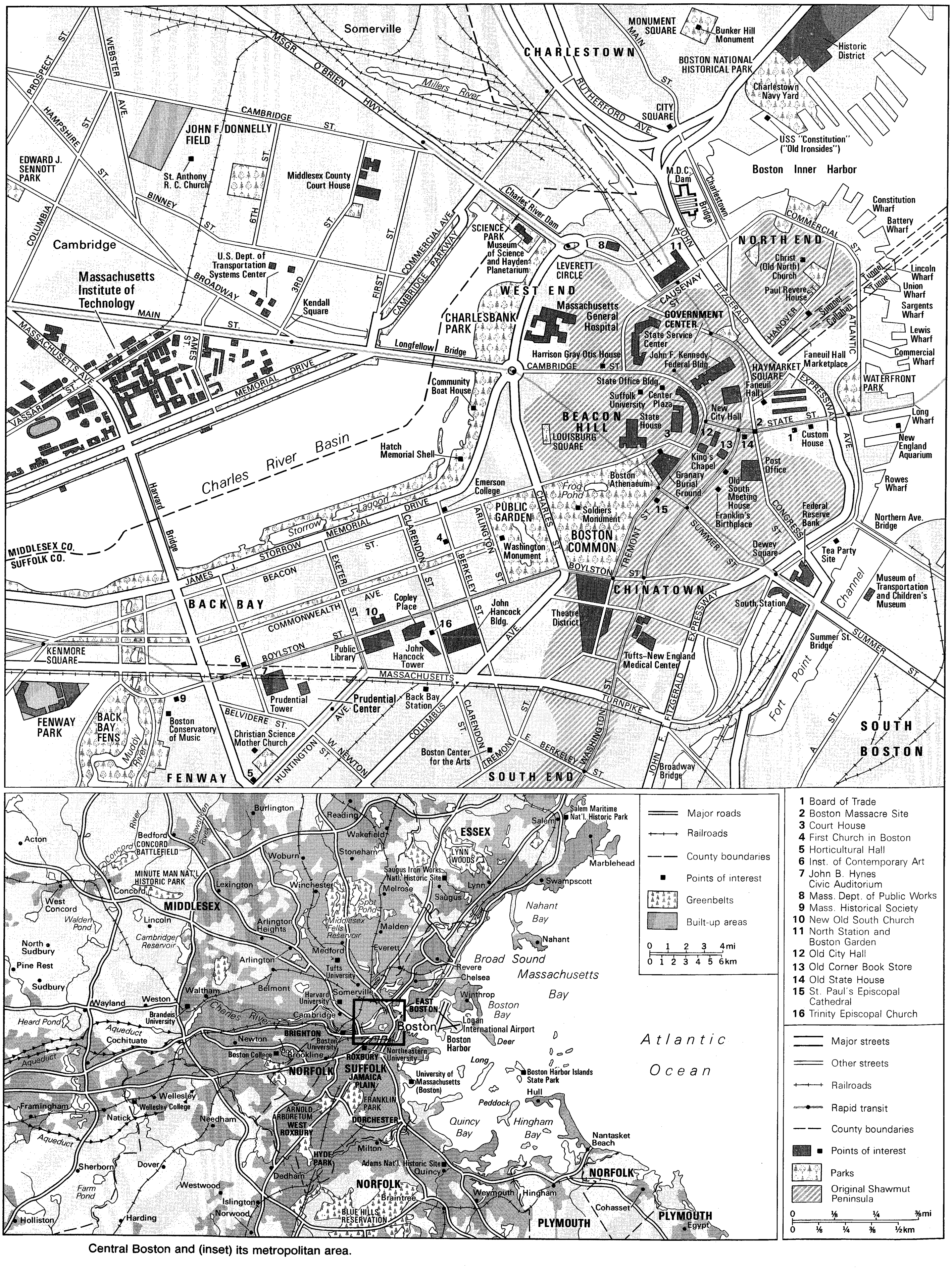
19th-century architecture

Bulfinch's designs. In the last years of the 18th century, when space became scarce, a series of major changes began to alter radically the physical lineaments of the area. In this period of expansion, the architect Charles Bulfinch (1763–1844), who for more than a quarter of a century was also the head of the town government, skillfully transformed an 18th-century English town into a 19th-century American city. Although many of Bulfinch's finest works have been destroyed, the central portion of the present State House (1795–98), above the Common on Beacon Hill, is his work. The construction of the State House on this site led to the conversion of the upland pastures of Beacon Hill into a handsome residential district that has survived with relatively little change. Between the State House and Charles Street are several streets, including famous Louisburg Square, filled with many houses by Bulfinch and other leading 19th-century architects. The area is protected by historic district legislation, and has been designated as the Beacon Hill Historic District.

As pressures of population in the 19th century caused a demand for more and more land, hills were cut down to fill in the coves. So much new land was created that the once water-ringed peninsula has become an indistinguishable part of the mainland. Fill on both sides of the narrow neck that connected the peninsula with the mainland created a new South End, a region that before the end of the century had lapsed into a slum. The waterfront was greatly extended, and the Back Bay was dammed (1818–21) to create tidal power for new mill sites. A causeway along the dam extended west from the Common to Sewell's Point, the present Kenmore Square in the Back Bay area, thus furnishing more direct communication with the mainland. The filling of Back Bay flats just west of Boston Common created land that in the 1830s was laid out as the Public Garden. This became a splendidly planted area with an artificial pond that is spanned by a miniature suspension bridge and traversed by swan-shaped excursion boats in the summer.

Fashioning of the Back Bay

The Back Bay mill basins never developed as their promoters had envisioned, partly because the introduction of railway lines through them in the 1830s hindered the



Central Boston and (inset) its metropolitan area.

flow of water. The area became a stinking nuisance, and the Massachusetts legislature in 1857 authorized its filling, thus creating a substantial amount of new land. The plan adopted provided for four new streets parallel to the Mill Dam (Beacon Street), to be intersected by cross streets. Commonwealth Avenue, which ran west from the Public Garden, was 200 feet (60 metres) wide with a park between its roadways, creating the atmosphere of a Parisian boulevard. Since there were no hills left to cut down, gravel had to be brought in by train from pits some miles away in Needham. By the end of the 19th century the Back Bay was completely filled and built up with houses that were subject to uniform height limits and setbacks. The region today presents a picture of American architecture that is as consistent for the second half of that century as Beacon Hill is for the first. Although many Back Bay houses have been converted to apartments, offices, schools, or other adaptive uses, the region has retained a good deal of its original character, and further changes are subject to architectural control.

Continuous park system. When the Back Bay was nearing completion during the 1880s, the American landscape architect Frederick Law Olmsted (1822–1903) developed an imaginative and large-scale design for the city's parks. It linked the Common, Public Garden, and Commonwealth Avenue with Franklin Park in West Roxbury by way of an open park known as the Fenway, which followed the Muddy River off to Brookline and Jamaica Plain. Included in this coordinated system was the Arnold Arboretum, a botanical outpost of Harvard University in Jamaica Plain. This led around the turn of the 20th century to a considerable western movement of institutions such as the Boston Symphony Orchestra, the New England Conservatory of Music, the Museum of Fine Arts, and the Harvard Medical School and hospitals. The completion in 1910 of a dam that kept the harbour tides out of the Charles River created a body of fresh water north of the Back Bay. The Charles River Basin, patterned after the Alster Basin in Hamburg, West Germany, remains one of the most handsome, distinctive, and popular features of Boston.

Suburbanization and annexations. The introduction, in the 1850s, of horsecar lines (later electrified) made it possible for many workers to move out of the confines of the city. Suddenly the adjacent towns were inundated. Roxbury, West Roxbury, Dorchester, Brighton, and Charlestown—once pleasant, independent communities—were absorbed into the city between 1855 and 1873. Everywhere, the line of demarcation between city and country was lost in a rapid and chaotic orgy of expansion and building.

Except for Hyde Park in 1912, no other towns have been brought into the city government of Boston in the 20th century. Although the city of Cambridge and the town of Brookline are independently governed, a casual visitor would hardly guess where Boston stops and they begin. The Charles River divides Boston from Cambridge, but both riverbanks appear to be part of one metropolis. Some regional cooperation is assured, however, through the Metropolitan District Commission, created in 1919 through the consolidation of the sewer and water districts and the park commission established earlier by the Massachusetts legislature.

Modern architecture. Until the mid-20th century, the low skyline of Boston was punctuated only by church steeples and by the tower of the Custom House, which as a federal building was not bound by the 125-foot height restriction that prevailed generally in the city. Modification of the building code, coincidental with a construction boom, brought about great changes in the 1960s, of which the first conspicuous example was the Prudential Center, with a 52-story tower on the site of train yards. The first major effort in urban renewal in Boston, initiated early in 1958, led to the wholesale demolition of the West End, the displacement of people, and the disruption of neighbourhoods to make way for the apartment towers of Charles River Park. Unfavourable reaction to such total and large-scale demolition of large areas without regard for the feelings of people led the mayor in 1960 to put an experienced lawyer at the head of the Boston Redevelopment Authority. During the next seven years the authority emphasized renewal rather than wholesale demolition and rebuilding. It endeavoured to enhance the traditional character of the city by bringing the best contemporary architecture and imaginative civic design to new projects. The Government Center, containing the new City Hall completed in 1968, in addition to federal, state, and private office buildings, thereby became a harmonious neighbour to Faneuil Hall and to other adjacent historic buildings. The creation of the Government Center stimulated considerable private construction in the vicinity in the 1960s. The John Hancock Mutual Life Insurance Company completed construction in 1976 of a tower that overtopped the Prudential Tower, while the Christian Science Church sponsored the reconstruction of large blocks of property near The Mother Church. Downtown waterfront renewal during the 1960s and 1970s along Atlantic Avenue and later developments on the site of the former Boston Naval Shipyard to the north have helped to open up the waterfront for commercial, residential, and recreational use and to preserve buildings of historical or architectural value. As seen from a distance, the profile of Boston is no longer low but is dominated by a series of towers from the waterfront to well inland.

Controls on urban renewal

The narrow and crowded streets of the central city are better suited for walking than driving, for Bostonians are incorrigible jaywalkers. The street markets around Faneuil Hall and in the North End are as essential a part of the city as ever, yet the terraced square of the Government Center provides a handsome setting for the modern City Hall that is a welcome great open space. The second-empire exterior of the old City Hall has been preserved by adapting its interior for revenue-producing use by offices, a bank, and a restaurant.

THE PEOPLE

Between 1800 and 1900 Boston changed from a relatively simple and homogeneous seaport of some 24,000 inhabitants, mainly of English extraction, to a polyglot city of more than 560,000, of whom about one-half were Irish and one-fifth were from the Canadian Maritime Provinces. After counting smaller groups of immigrants from Russia,

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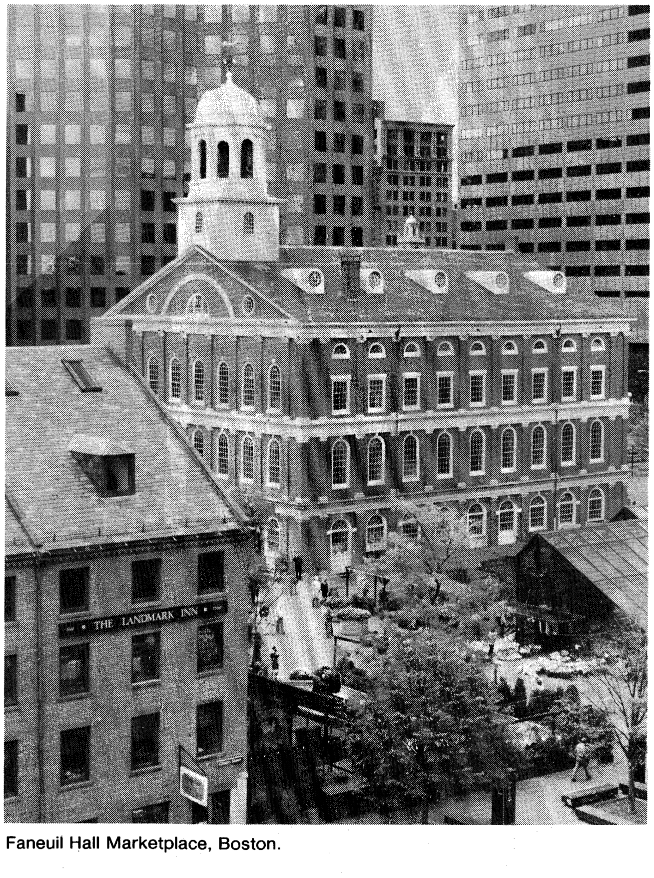

Faneuil Hall Marketplace, Boston.

Italy, Germany, and other European countries, as well as American-born blacks, less than 11 percent of the city's population was of traditional New England origin. The burgeoning of Boston's population in the 19th century was due to seemingly endless waves of immigrants that increased the total by at least one-fourth in each decade from 1810 to 1900; during five decades the increase was more than 40 percent. The greatest increases were 52 percent in the 1830s and 47 percent in the 1840s. Many of the new arrivals, who brought with them experience in trades and crafts, though disembarking at Boston, rapidly moved westward to the expanding frontier, where their skills were in demand. Those with only unskilled labour to offer remained in Boston, crowded into urban slums. As earlier arrivals prospered and moved elsewhere, often outside the city limits, their places were taken by more unskilled immigrants. By the middle of the 19th century, Irish immigrants had taken over the North End, but in the last third of the century it became the centre of the first exclusively eastern European Jewish community in Boston. The residents are now predominantly of Italian descent. Similar changes have occurred in many parts of the city.

Inundation by immigration

In the 20th century the greatly reduced immigration has helped to slow the population increase. Through migration from the South, the black population of Boston has risen, the majority crowded into the decaying Roxbury section.

THE ECONOMY

Finance. Banking, insurance, investment management, and other financial and business-service activities are the chief elements in the economic life of the city. The Federal Reserve Bank of the New England district is located in Boston.

Industry. The chief industries are food processing, clothing manufacturing, printing and publishing, and the fabrication of machinery, transport equipment, and instruments. The greater part of the research-and-development activity that burgeoned after World War II is based in the metropolitan area outside the city limits.

Although Boston continues as a centre of retail trade, much wholesale activity has moved outside the city. The great number of universities and hospitals make higher education and health a significant element in the economic life of Boston. Although only one-fifth of the people in the metropolitan area live in Boston, the city produces almost half of the metropolitan area's goods. A large share of the jobs in Boston are filled by commuters from the suburban area.

Transportation. Public transportation is maintained by the Massachusetts Bay Transportation Authority through a network of subway, elevated, and surface lines. The subway system, begun in 1897, was the first in the country, and, though it has been designated as a historic place, it still works to a useful purpose.

Railroad passenger service to the metropolitan area, once excellent, has badly deteriorated through competition from private automobiles and buses. Such vehicles increasingly choke the city, especially the narrow and winding downtown streets laid out in colonial times. Logan International Airport in East Boston is a busy centre for European as well as domestic flights.

ADMINISTRATIVE AND SOCIAL CONDITIONS

Government. Boston is governed by a mayor, who is elected for a four-year term, and by a city council of nine members. The government of the Commonwealth of Massachusetts is centred in the State House at the top of Beacon Hill and in other buildings adjacent to the Government Center.

Other federal offices occupy an earlier federal building and the Custom House. Boston is the headquarters of the 1st Naval District, but the Boston Naval Shipyard was closed in 1974.

Education and health. Universities, colleges, and schools of every kind fill several pages in the classified section of the Boston telephone directory. Boston University (founded 1869), Northeastern University (1898), Suffolk University (1906), and the Boston campus of the University of Massachusetts, as well as Simmons (1899), Emmanuel (1919), and Emerson (1880) colleges, are based within the city, as are the Harvard University (1636) faculties of medicine, dentistry, public health, and business administration.

The rest of Harvard, Radcliffe College (1879), and the Massachusetts Institute of Technology (1861) are just across the Charles River in Cambridge. Boston College (1863), a Jesuit institution that is a university in everything except name, is in Chestnut Hill, only a step beyond the city limits. Tufts University (1852), although based in Medford, has its medical school in Boston. Massachusetts General, the oldest of the hospitals, has been joined, among others, by the Beth Israel, Children's, Deaconess, and Brigham and Women's hospitals and the Tufts–New England Medical Center. There are few aspects of higher education or health that are not dealt with somewhere within or close to the city.

Hospitals

CULTURAL LIFE

Boston has a rich and varied cultural life. The city has a repertory theatre group and resident ballet and opera companies; in addition, numerous road companies visit the city. Boston and its environs offer opportunities for all kinds of participatory or spectator sports, including several major-league teams.

The Boston Symphony Orchestra, which was founded in 1881 and performs in Symphony Hall, is one of the foremost orchestras in the world and the chief musical attraction of Boston. The New England Conservatory of Music is nearby, as are other music schools. Chamber music abounds, as does an interest in early music encouraged by the presence of several harpsichord and other instrument makers in the region. The summer Esplanade Concerts at the Hatch Memorial Shell by the Charles River Basin are one of the few outdoor events that have thrived consistently over a long period. Most of the concerts are given by the Boston Pops Orchestra. Overall, however, the cultural life of the city depends more upon colleges, museums, libraries, and small private groups than upon a public proliferation of the performing arts.

The Museum of Fine Arts, a major world gallery incorporated in 1870, preserves and exhibits Oriental, Egyptian, and classical collections of international distinction, as well as important examples of paintings, prints, textiles, and the decorative arts. The Isabella Stewart Gardner Museum in the Fenway is rich in Italian Renaissance painting. Since its arrangement has, under terms of its deed to the city, remained unchanged since the death in 1924 of its creator, Mrs. John Lowell Gardner, in whose home the collection is housed, it is an important monument as well in the history of taste. Across the river at Harvard, there are also a fine arts museum and several museums devoted to science and natural history. The Museum of Science at Science Park, overlooking the Charles River Basin, and the Children's Museum at Museum Wharf are aimed at the instruction of the young.

Museums and galleries

Boston is equally rich in libraries. The Boston Public Library (1854) was the first major instance in the United States of a tax-supported free library. It has been housed since 1895 in a building designed by the architectural firm of McKim, Mead, and White and is a masterpiece of 19th-century American architecture. Its interior decor, including murals by the Frenchman Pierre Puvis de Chavannes (1824–98) and the American John Singer Sargent (1856–1925), is a striking feature. The library, with a fine collection, carries out many of its functions through branches spread throughout the city. The Boston Athenaeum, a proprietary library founded in 1807, the Massachusetts Historical Society, the New England Historic Genealogical Society, the State Library, and the libraries of the colleges and universities provide remarkable resources, while, across the Charles, Harvard has one of the greatest libraries for scholarship in the world.

College crews row on the Charles River, and small boats sail on the wide basin between the Back Bay and the Cambridge shore. Little recreational use has been made of the harbour, however, for in the 19th century Boston turned its back on the sea. The Boston Redevelopment Authority has worked to reverse this trend. The downtown waterfront area contains the New England Aquarium, and a section of the A.G. Charlestown Navy Yard has been made part of the Boston National Historic Park. The fine park system designed by Olmsted has suffered from the encroachments of automobiles, except in the Arnold Arboretum in Jamaica Plain, from which they are excluded.

One of the principal forms of recreation in Boston is just looking at the city. Many thousands of visitors come annually to Boston because of its numerous historic sites and the architectural character of the city. Boston has preserved some colonial buildings and an extraordinary number of fine 19th-century structures, yet in the city and its surroundings may be seen a great variety of the works of the principal American and European architects of the second half of the 20th century.

History

THE COLONIAL PERIOD

Settlement and growth. Boston was settled in 1630 by Puritan Englishmen of the Massachusetts Bay Company, who, for religious reasons, put the Atlantic Ocean between themselves and the Church of England. Ostensibly founded for commercial reasons, the Massachusetts Bay Company, under its governor, John Winthrop (1588–1649), brought its charter—which it regarded as authorization to set up a self-governing settlement in the New England wilderness—along to the New World. The new town was named for Boston in Lincolnshire, the former home of many of the immigrants.

Through necessity rather than choice, New Englanders turned to the sea for a livelihood and became shipbuilders, merchants, seamen, and fishermen because there was little else to do. The Shawmut Peninsula, upon which Boston was settled, was an ideal setting for a seaport. It was described in 1634 by William Wood in his *New England Prospects* as "fittest for such as can Trade into England, for such commodities as the Country wants, being the chiefe place for shipping and Merchandize." With the triumph of the Puritan Party in England in 1648, people moved freely between New England and the homeland, and close ties of family and trade linked Boston and London. By the end of the 17th century, Boston's fleet of seagoing vessels was exceeded only by those of London and Bristol in the English-speaking world. Boston held its place as the largest town in British North America until the middle of the 18th century, when it fell behind the faster growing ports of Philadelphia and New York City.

Political life and revolutionary activity. During its first 50 years Boston was a homogeneous, self-governing Puritan community in which the leaders of the Massachusetts Bay Company ruled as they saw fit. The three Puritan churches, established on the congregational principle, accounted for almost all the organized religion in Boston. Religious dissidents were banished, and some Quakers who persisted in returning were hanged for their pains. Although the increasing prosperity of the colonial merchants made London quite aware of Massachusetts Bay, steps to assert royal authority there were taken only near the end of Charles II's reign, in the 1680s. The company's charter was declared null and void in 1684. In 1686, with the arrival of Sir Edmund Andros as the first royal governor of the Province of Massachusetts Bay, the authority of the crown was established in Boston itself. With this change, the Church of England first came to the town, and the Puritan isolation was over. Although the Congregational clergy, particularly the voluble father–son combination of Increase (1639–1723) and Cotton (1663–1728) Mather, still made themselves heard, the lines of authority had altered.

Protests against British rule

Boston never proved wholly docile. When word of the "Glorious Revolution" of 1688 reached them, the citizens on April 18, 1689, dumped Andros out of office and imprisoned him. The memory of the autonomous first half-century lingered. As London endeavoured to enforce navigation laws and gain revenue from the Boston trade at the expense of the colonies, the aggrieved inhabitants indulged in what they felt to be justified resistance against unlawful authority. After passage of the Stamp Act by Parliament in 1765, disaffection grew, and the governor's house was stormed and gutted, an act that destroyed many irreplaceable records of the colony's history. The Boston Massacre of 1770, in which British troops fired on a crowd of civilian hecklers and killed several persons, and the Boston Tea Party of 1773, in which colonists disguised as Indians dumped three shiploads of tea into Boston Harbor, were key incidents in measuring popular sentiment prior to the Revolution.

With the confrontations and exchange of shots at Lexington and Concord on April 19, 1775, the die was cast. When George Washington's army besieged the British in Boston during the following winter, normal life in the town was suspended. On March 17, 1776, impelled by Washington's artillery positioned on Dorchester Heights, British troops and officials left. They were accompanied by loyal supporters of the crown, including a number of the principal merchants. A constitution was framed in 1780, and John Hancock was elected the first governor of the Commonwealth of Massachusetts.

BOSTON AFTER 1776

Adjustment to independence. New men risen from obscurity by way of political or military service or ventures at sea during the Revolution quickly filled the places left vacant by departed Loyalists. Independence gravely imperilled Boston's maritime trade, for, at the close of the Revolution, Boston merchants automatically became foreigners in the ports of the British Empire. Thus, survival depended on finding new channels of trade. The crisis was solved by sending ships to distant and hitherto unfamiliar ports. The development of the China trade and other new routes, such as those to India, raised Boston to greater prosperity than ever before.

Throughout the first half of the 19th century, maritime commerce produced substantial fortunes, which were supplemented by others achieved in mercantile and manufacturing pursuits without recourse to the sea. Bostonians in the 1810s began to establish textile mills, first at Waltham in 1813 and then in new towns along the Merrimack River, where waterpower was plentiful. The advent of railroads in the 1830s brought these once-distant towns suddenly closer. The burgeoning of Boston's population was due not only to maritime commerce and manufacturing but also to the unanticipated arrival of immigrants from Europe in such numbers that the city grew more than 20-fold during the 19th century. By 1822 the traditional form of government, in which a board of selectmen administered the decisions reached by the vote of all citizens at an annual town meeting, had become unmanageable, and a city charter was obtained from the legislature.

The era of culture. The 19th century also saw Boston assume its focal position in the religious and educational life of the new nation. The rapid and large-scale infusion of immigrant groups and the loss of dominance by the Congregational descendants of the Puritan settlers were major factors in the change. When it was built in 1799, the Church of the Holy Cross was the only Roman Catholic church in town. In 1808 its founder, Jean-Louis Lefebvre de Cheverus, a refugee from the French Revolution, was created the first bishop of Boston, and in 1875, when the see was raised to the dignity of an archdiocese, there were 28 Catholic parishes.

Religious diversity

The religious climate of Protestant Boston also changed dramatically. The Anglican King's Chapel, whose Loyalist rector had left with the British in 1776, became Unitarian in 1787, though not until the first quarter of the 19th century did Unitarian doctrine produce permanent cleavage and splinter organizations in many Congregational churches. The creation in 1825 of the American Unitarian Association, in which William Ellery Channing (1780–1842) took the leading part, transformed into a separate denomination what had previously been the liberal wing of Congregationalism. From "Channing Unitarianism" it was only a step to the philosophy of Transcendentalism, which affected a great part of the region's artistic output and thought for much of the century; to the abolitionist movement championed by the outspoken writer and editor William Lloyd Garrison (1805–79) and others; and to liberal high-mindedness in social causes, which preoccupied many 19th-century Bostonians at a time when others were simply making money with great Yankee diligence. In the second half of the 19th century, the city was also the founding place of Christian Science by Mary Baker Eddy (1821–1910). It remains the site of The Mother Church and the headquarters of this faith, which grew into an international Christian denomination.

Educational and cultural institutions had a similarly rapid growth. Across the Charles River in New Towne, later Cambridge, a college had been founded in 1636 to provide the infant colony with religious scholars and ministers. It was named for the Charlestown minister John Harvard (1607–38), who bequeathed his library to the institution in 1638, and was the sole college in the area until the third quarter of the 19th century. Though Harvard retained the most prestigious position throughout the century, a host of other major institutions of higher learning were founded, and Boston became synonymous nationally with scholarship and cultural refinement. It became also the mecca for persons—from abroad or from the "less civilized" parts of the nation outside New England—who sought these qualities amid the bustling commercialism and rambunctious growth that characterized much of 19th-century America.

Financial growth. The Civil War, in which a majority of Bostonians strongly supported the Union, put an end to shipping as a major consideration in the life of Boston. Shipowners and merchants invested thenceforth in manufacturing, in railroads, and in the development of the rapidly expanding frontier. Banking and investment, with the interplay that they implied with various forms of manufacturing and of business, superseded maritime commerce as the principal occupation of Boston in the second half of the 19th century. In the early 20th century, however, the business horizons contracted. Though large sums of money continued to be invested outside of New England, fewer distant companies were controlled from Boston. The city's financial capital was as strong, but the rising strength of New York City and Chicago and of the developing states of the American West gradually reduced the proportion of capital that Boston could muster. Nevertheless, Boston's financial-management firms showed a skill in investment that caused them to be well regarded in other parts of the country. This led eventually to a major growth of those Boston companies that administered mutual investment funds. Thus, the "prudent man," whether in a private trustee's office or an investment company, has survived as a Boston asset, whereas the textile mills and railroads have proved to be less permanent. The textile industry passed into crisis in the 1920s, and the industrial cities on the Merrimack River, created by Boston investment, entered on decades of hardship. In the following years some mills went out of business entirely, while others moved to the South in search of cheaper labour and raw materials.

Problems in the early 20th century

The first half of the 20th century in Boston was a divisive and unhappy period, for, in addition to corruption in local politics, it included two world wars separated by Prohibition and the Great Depression. After World War I the increase of automobiles led more and more Bostonians to move outside the city limits, with consequent detriment to older central residential districts. Many of the defense industries that sprang up after World War II, based on the imagination of scientists from Harvard, the Massachusetts Institute of Technology, and other centres, were located outside the city limits.

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The House of Bourbon

The House of Bourbon (Spanish Borbón, Italian Borbone) is one of the greatest of the sovereign dynasties of Europe. It provided reigning kings of France from 1589 to 1792 and from 1814 to 1830, after which another Bourbon reigned as king of the French until 1848; kings or queens of Spain from 1700 to 1808, from 1814 to 1868, from 1874 to 1931, and since 1975; dukes of Parma from 1731 to 1735, from 1748 to 1802, and from 1847 to 1859; kings of Naples and of Sicily from 1734 to 1808 and of the Two Sicilies from 1816 to 1860; kings of Etruria from 1801 to 1807; and ducal sovereigns of Lucca from 1815 to 1847. The present article attempts a rapid survey of the dynasty as a whole, relying mainly on genealogical tables to display necessary details. In these tables the names and titles of sovereigns are mostly anglicized, but those of other persons are mostly given in the original form, except where princesses, having married into another country, are better known under that country's name for them. The tables also omit perforce the Bourbon bastards, whose multitude lends some colour to the popular notion that the "Bourbon nose" (larger and more prominent than the normal aquiline) betokens a "Bourbon temperament" or enormous appetite for sexual intercourse.

Origins. The House of Bourbon is a branch of the House of Capet, which constituted the so-called third race of France's kings. King Louis IX, a Capetian of the "direct line," was the ancestor of all the Bourbons through his sixth son, Robert, comte de Clermont (see Table 1). When the "direct line" died out in 1328, the House of Valois, genealogically senior to the Bourbons, prevented the latter from accession to the French crown until 1589. The Valois, however, established the so-called Salic Law of Succession, under which the crown passed through males according to primogeniture, not through females. On this principle, the senior Bourbon became the rightful king of France on the extinction of the legitimate male line of the Valois.

The first ducs de Bourbon

Robert de Clermont had married the heiress of the lordship of Bourbon (Bourbon-l'Archambault, in the modern *département* of Allier). This lordship was made a duchy for his son Louis I in 1327 and so gave its name to the dynasty. From this duchy, the nucleus of the fu-

Table 1: The Bourbon Descent from Louis IX of France

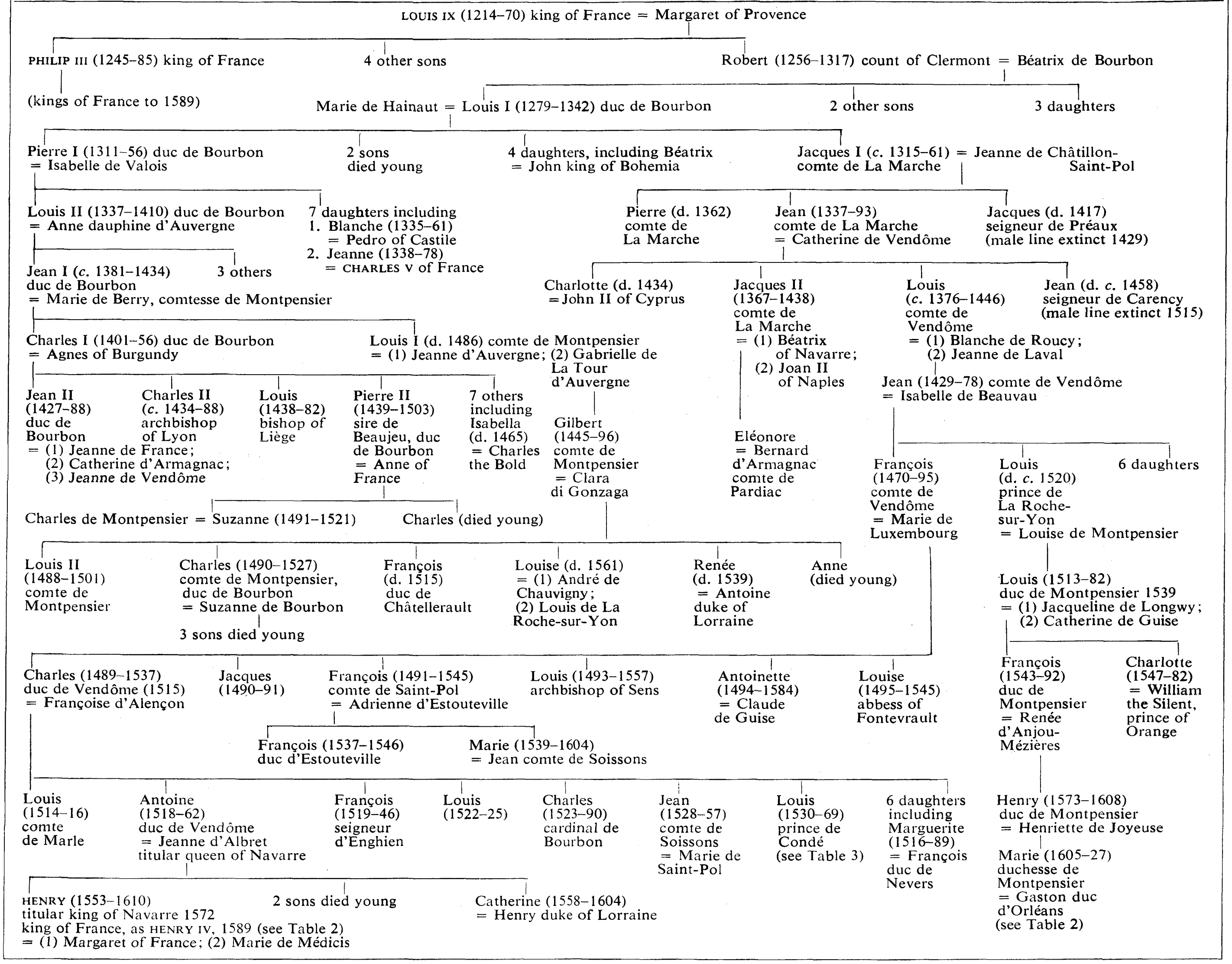
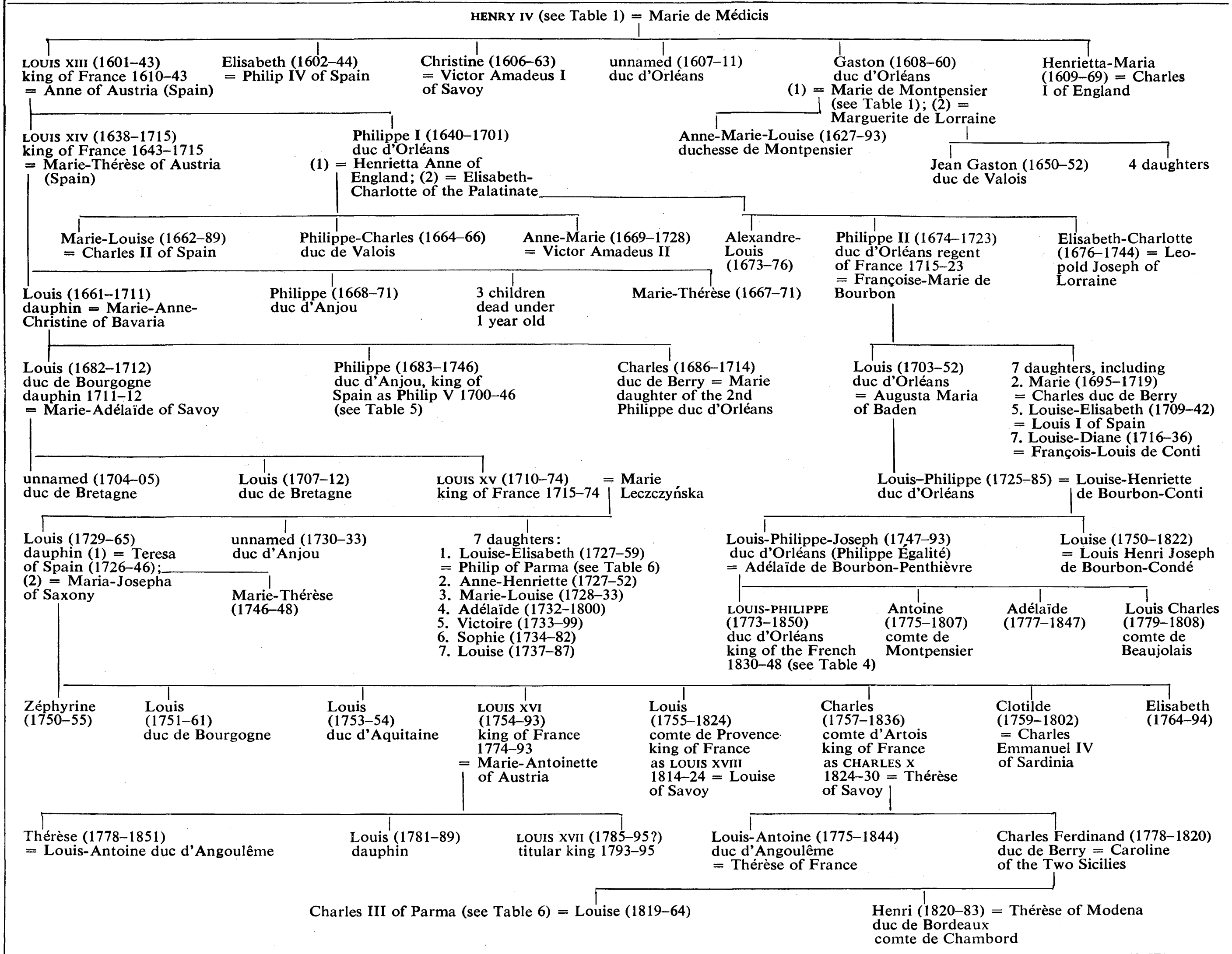


Table 2: The Bourbon Kings of France, with the House of Orléans to 1830



ture province of Bourbonnais, the elder Bourbons, mainly through marriages, expanded their territory southeastward and southward. On their western frontier, meanwhile, the countship of La Marche (acquired by Louis I in 1322 in exchange for Clermont) was held from 1327 by a junior line of Louis I's descendants, who soon added the distant countship of Vendôme to their holdings.

The title of duc de Bourbon passed in 1503 to Charles de Bourbon-Montpensier, who was to become famous as constable of France. His later treason led to the confiscation of his lands by the French crown in the year of his death, 1527. Headship of the House of Bourbon then passed to the line of La Marche-Vendôme (see again Table 1).

The line of La Marche-Vendôme had been subdivided since the end of the 15th century between a senior line, that of Vendôme (with ducal rank from 1515 onward), and a junior one, that of La Roche-sur-Yon. The latter line obtained Montpensier from the constable's forfeited heritage (with ducal rank from 1539).

The French kingdom

Antoine de Bourbon, duc de Vendôme and head of the House of Bourbon from 1537, became titular king-consort of Navarre in 1555 through his marriage in 1548 to Jeanne d'Albret. The son of that marriage, titular king of Navarre in succession to his mother from 1572, became king of France, as Henry IV, on the death of the last Valois king in 1589. From Henry IV descend all the Bourbon sovereigns (see Tables 2 and 4–7). The great House of Condé, with its ramifications of Soissons and of Conti (see Table 3), was descended from Louis, prince de Condé, one of Henry IV's uncles.

The Bourbon sovereignties. Henry IV's heirs were kings of France uninterruptedly from 1610 to 1792, when the monarchy was "suspended" during the first Revolution. Most illustrious among them was Louis XIV, who brought absolute monarchy to its zenith in western Europe. Table 2 displays (1) the senior line of Bourbon, or House of France, as descended from Henry IV, the succession running directly down to Louis XVII (who was titular king only and never reigned), and thereafter being resumed in the latter's uncle, Louis XVIII (brought to the throne in 1814 and again in 1815), Charles X, and the pretender Henry V (the comte de Chambord); (2) the connection, through Louis XIV's grandson Philippe, duc d'Anjou, of the Spanish Bourbons with the House of France; and (3) the descent of the collateral House of Orléans from Louis XIV's brother Philippe I to Louis-Philippe, king of the French. Louis-Philippe's descendants, including not only the potential pretenders to the French succession but also the Bourbon descendants of the heiress of the last emperor of Brazil, are shown in Table 4. The last-mentioned princes constituted the House of Bourbon-Brazil, or of Orléans-Braganza, which is not to be confused with the House of Borbón-Braganza, a Spanish branch originating in the Portuguese marriage of the infante Don Gabriel (a son of Charles III of Spain; see Table 5).

Accession in Spain

The Bourbon accession to Spain came about partly because the descendants of Louis XIV's consort, the Spanish infanta Marie-Thérèse, were in 1700 the closest surviving relatives of the childless Charles II of Spain (see HABSBURG; SPAIN); and partly because, although at her

Table 3: Condé, Conti, and Soissons

Louis I (1530–69) prince de Condé (see Table 1)
= (1) Eléonore de Roye; (2) Françoise d'Orléans-Longueville

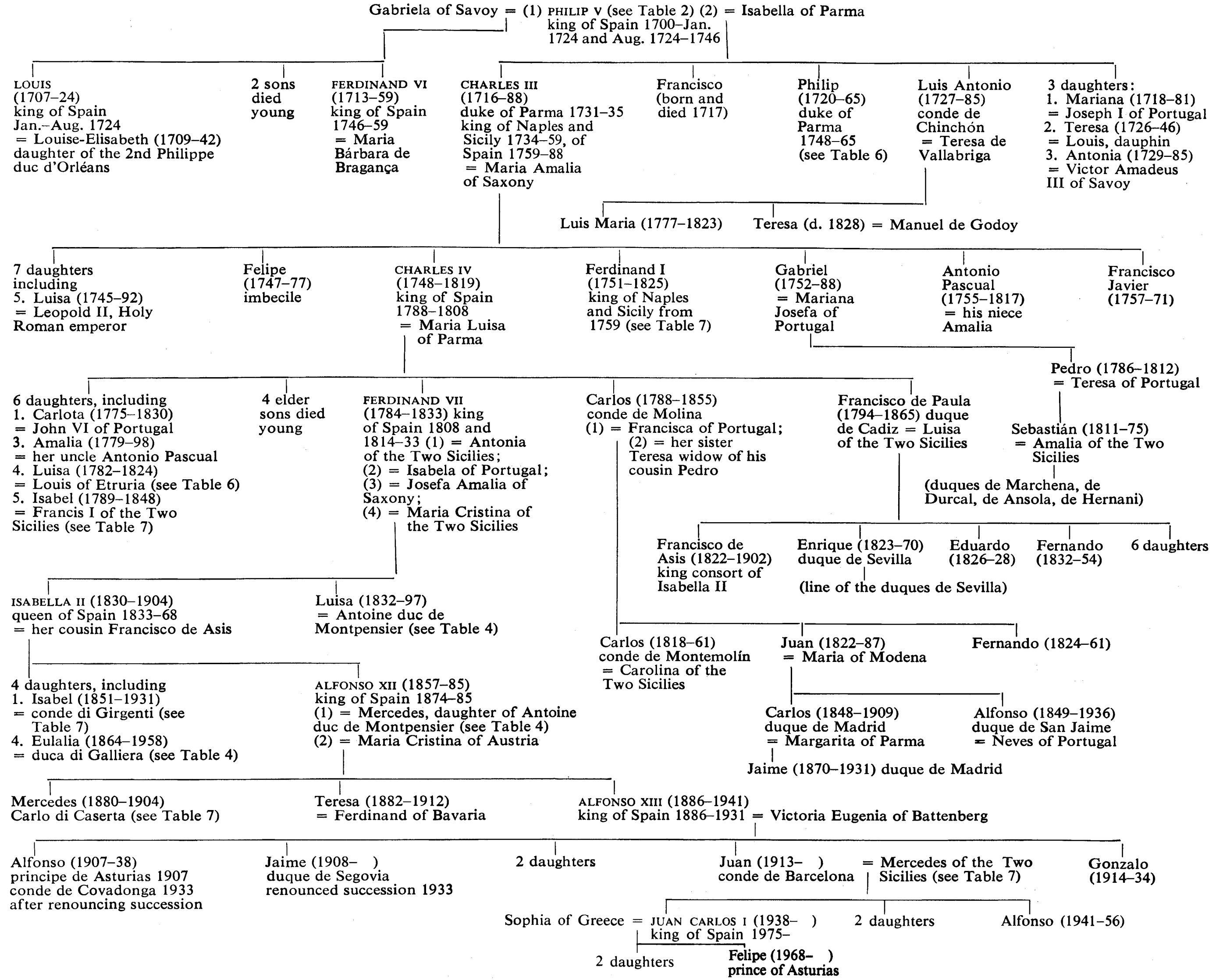
- Henry I (1552–88) prince de Condé = (1) Marie de Clèves-Nevers; (2) Charlotte de La Trémoille
  - Catherine (1574–95)
  - Eléonore (1587–1619) = Philip William of Orange-Nassau
  - Henry II (1588–1646) prince de Condé = Charlotte de Montmorency
    - Anne-Geneviève (1619–79) = Henri d'Orléans duc de Longueville
    - Louis II (1621–86) duc d'Enghien till 1646, then prince de Condé = Claire-Clémence de Maillé-Brézé
      - Henri-Jules (1643–1709) prince de Condé = Anne of Bavaria daughter of Edward of the Palatinate
        - Henri (1657–70)
        - Louis III (1668–1710) duc de Bourbon = Louise-Françoise legitimated daughter of Louis XIV
          - Louis-Henri (1692–1740) duc de Bourbon = (1) Marie-Anne de Bourbon-Conti; (2) Charlotte of Hesse-Rheinfels
            - Louis-Joseph (1736–1818) prince de Condé = Godefride de Rohan-Soubise
              - Louis-Henri-Joseph (1756–1830) duc de Bourbon = Louise d'Orléans (1750–1822; see Table 2)
                - Louis-Antoine (1772–1804) duc d'Enghien
              - Louise-Adélaïde (1757–1824) Benedictine nun
          - 2 other sons: 1. Charles (1700–60) comte de Charolais; 2. Louis (1707–71) comte de Clermont
          - 6 daughters including 2. Louise-Elisabeth = Louis-Armand II prince de Conti
        - 2 sons died in infancy
        - 6 daughters including
          1. Marie-Thérèse (1666–1732) = François-Louis de Bourbon-Conti
          4. Anne-Louise-Bénédicte (1676–1753) = Louis-Auguste duc du Maine legitimated son of Louis XIV
          5. Marie-Anne (1678–1718) = Louis-Joseph duc de Vendôme
      - 2 sons died in infancy
    - Armand (1629–66) prince de Conti = Anne-Marie Martinozzi niece of Cardinal Mazarin
      - 1 son died in infancy 1658
      - Louis-Armand I (1661–85) prince de Conti = Marie-Anne legitimated daughter of Louis XIV
      - François-Louis (1664–1709) comte de La Marche 1677–85, then prince de Conti = Marie-Thérèse de Bourbon-Condé
        - Louis-Armand II (1695–1727) prince de Conti = Louise-Elisabeth de Bourbon-Condé
          - Louis-François (1717–76) prince de Conti = Louise-Diane (1716–36) daughter of the regent duc d'Orléans (see Table 2)
            - Louis-François-Joseph (1734–1814) prince de Conti = Maria Fortunata d'Este
          - 3 sons died in infancy
          - Louise-Henriette (1726–59) = Louis-Philippe (1725–85) duc d'Orléans (see Table 2)
        - 3 sons died in infancy and 3 daughters, including 1. Marie-Anne (1689–1720) = Louis-Henri duc de Bourbon
- François (1558–1614) prince de Conti = (1) Jeanne de Coësme; (2) Marguerite de Guise
- Charles (1562–94) cardinal de Vendôme
- 5 others
- Charles (1566–1612) comte de Soissons = Anne de Montafié
  - Louis (1604–41) comte de Soissons
  - 3 daughters:
    1. Louise (1603–37) = Henri d'Orléans duc de Longueville;
    2. Marie (1606–92) = Thomas de Savoie prince de Carignan;
    3. Charlotte-Anne (1608–23)
- 2 sons died young

Table 4: The House of Orléans from 1830

LOUIS-PHILIPPE (see Table 2) king of the French 1830–1848 = Marie-Amélie of the Two Sicilies (see Table 7)

- Ferdinand (1810–42) duc d'Orléans 1830–42 = Helena of Mecklenburg-Schwerin
  - Louis-Philippe-Albert (1838–94) comte de Paris = Isabel daughter of Antoine duc de Montpensier
    - Louis-Philippe-Robert (1869–1926) duc d'Orléans = Maria of Austria
    - 4 daughters
    - Ferdinand (1884–1924) duc de Montpensier
  - Robert (1840–1910) duc de Chartres = Françoise de Joinville
    - 2 daughters
    - Robert (1866–83)
    - Henri (1867–1901)
    - Jean (1874–1940) duc de Guise = Isabelle daughter of comte de Paris
      - Henri (1908–) comte de Paris = Isabelle of Orléans-Braganza
        - Henri (1933–) comte de Clermont = Marie-Thérèse of Württemberg
          - François (1961–)
          - Jean (1965–)
          - 2 daughters
        - 4 other sons
        - 6 daughters
      - 3 daughters
- Louis (1814–96) duc de Nemours = Victoria of Saxe-Coburg
  - Gaston (1842–1922) comte d'Eu = Isabela (1846–1921) heiress of Pedro II of Brazil
    - Pedro (1875–1940) renounced succession 1908
      - Henri comte de Paris = Isabelle (1911–)
      - 4 others
    - Luis (1878–1920) = Pia, daughter of Alfonso di Caserta (see Table 7)
      - Pedro Enrique (1909–) = Maria of Bavaria
        - 6 sons
    - Antonio (1881–1918)
  - Ferdinand (1844–1910) duc d'Alençon = Sophia of Bavaria
    - 1 daughter
    - Emmanuel (1872–1931) duc de Vendôme, duc d'Alençon = Henriette of Belgium
      - 3 daughters
      - Charles Philippe (1905– duc de Nemours = Marguerite Watson
- François (1818–1900) prince de Joinville = Francesca of Brazil
  - Françoise (1844–1925) = duc de Chartres
  - Pierre (1845–1919) duc de Penthièvre
- Charles (1820–29) duc de Penthièvre
- Henri (1822–97) duc d'Aumale = Carolina daughter of principe di Salerno (see Table 7)
  - Louis (1845–66) prince de Condé
  - François (1854–72) duc de Guise
- Antoine (1824–90) duc de Montpensier = infanta Luisa (see Table 5)
  - Fernando (1859–73)
  - Antonio (1866–1930) duca di Galliera = infanta Eulalia (see Table 5)
    - 2 sons
  - Luis (1867–74)
  - 3 daughters, including
    2. Isabel (1848–1919) = comte de Paris
    3. Mercedes (1860–78) = Alfonso XII of Spain
- 3 daughters:
  1. Louise (1812–50) = Leopold I of the Belgians
  2. Christine (1813–39) = Frederick William of Württemberg
  3. Clementine (1817–1907) = Augustus of Saxe-Coburg

Table 5: The Spanish Bourbons



marriage the Infanta had renounced her Spanish rights, Charles by his testament named one of her descendants as his successor. Since the other powers, however, would not have tolerated the union of the Spanish kingdom with the French, Charles named neither Louis XIV's heir apparent nor the latter's eldest son but, rather, the second of Louis XIV's grandsons, namely Philippe duc d'Anjou, who became king of Spain as Philip V. After the War of the Spanish Succession, the Peace of Utrecht (1713) left Philip in possession of Spain and Spanish America but obliged him to renounce any natural right that he or his descendants might have to France. Table 5 shows how the Spanish Bourbons all descend from Philip V; it also shows the origin of the branches of Parma and of Naples-Sicily.

Parma and Naples-Sicily

The infante Don Carlos, the future Charles III of Spain, was the founder of the Bourbon fortunes in Italy. The eldest son of Philip V's second marriage, he became duke of Parma in 1731 by right of his mother, heiress of the last Farnese dukes; and in 1734, during the War of the Polish Succession, he conquered the Kingdom of Naples-Sicily (Kingdom of the Two Sicilies) for himself. Though the settlement of 1735–38 obliged him to renounce Parma in order to win international recognition as king of Naples-Sicily, Parma was eventually secured for his brother Philip (Don Felipe) under the Treaty of Aix-la-Chapelle in 1748—with the proviso, however, that he and his heirs should renounce it in the event that they succeeded to Naples-Sicily or to Spain. Finally, when Don Carlos became king of Spain as Charles III in 1759, he resigned Naples-Sicily to his third son Ferdinand on the express condition that that kingdom and Spain should never be united under one sovereign.

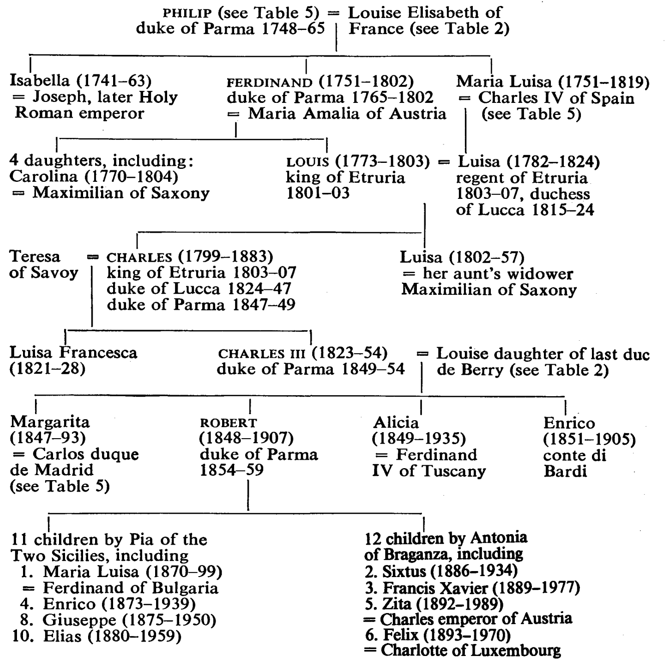
Table 6 shows the descendants of Philip of Parma, Table 7 those of Ferdinand I of Naples-Sicily.

Etruria and Lucca

The Kingdom of Etruria (1801–07) was a contrivance of the Napoleonic period. Devised by the French for the House of Bourbon-Parma in compensation for the impending annexation of Parma to France at a time when France still needed the goodwill of the Spanish Bourbons, it was dissolved as soon as Napoleon was ready to depose the latter. The Bourbon Duchy of Lucca (1815–47), on the other hand, was a creation of the Congress of Vienna: having assigned Parma to Napoleon's estranged consort Marie-Louise for her lifetime, the Congress had to find some alternative compensation for the still-dispossessed Bourbons. The Treaty of Paris of 1817, however, prescribed that on Marie-Louise's death Parma should revert to the Bourbons, who in 1847 renounced Lucca to the Habsburgs of Tuscany nine weeks before succeeding her (see again Table 6).

In France, the senior or "legitimate" line of the Bourbons, restored to sovereignty in France after the Napoleonic Wars, was deposed at the Revolution of 1830. The House of Orléans, which took the legitimate line's place, was in turn deposed in the Revolution of 1848. The Bourbons of Parma and of the Two Sicilies were dethroned in 1859–60,

Table 6: The Bourbons of Parma



in the course of the unification of Italy under the House of Savoy. The Spanish Bourbons, after many disturbances in the 19th century, lost their sovereignty in 1931; but the Law of Succession promulgated in Spain in 1947 and General Franco's subsequent choice of Don Juan Carlos as his successor (see again Table 5) resulted in the reign of Juan Carlos I, beginning in 1975.

Solidarity and discord. The accession of the duc d'Anjou to Spain would never have been secured without the resolute support of his grandfather, the French king; and similarly the Bourbon sovereignties in Italy owed their establishment chiefly to the Bourbon power in Spain. Dynastic harmony between France and Spain, however, was momentarily suspended in 1718–20, when France took part in the War of the Quadruple Alliance against Spain—for reasons arising in part from the internal affairs of the House of Bourbon. A series of sudden deaths in the French Royal House between 1704 and 1714 had produced a situation in which, on Louis XIV's death in 1715, no one but a five-year-old child, Louis XV, stood before Philip V of Spain in the natural line of succession of France; and Philip, though he had renounced that succession, still felt himself better entitled, as the child's uncle, to exercise the regency in France than the child's cousin twice removed, Philippe duc d'Orléans, against whom Spanish agents promoted a plot. The marriage (1722) of the Spanish king's son to a daughter of the French regent sealed the reconciliation.

Treaty of the Escorial

In 1733 the Treaty of the Escorial pledged the French and the Spanish Bourbons to collaborate with each other notwithstanding any previous obligations. This treaty and the similarly conceived Treaty of Fontainebleau (1743) are sometimes called the "First" and the "Second Family Compact"; and the term Family Compact, or Pacte de Famille, was actually used in a third treaty, signed in Paris in 1761, during the Seven Years' War. By this last treaty France and Spain not only guaranteed one another against all enemies but also promised like protection to the Bourbon states in Italy in the event of their acceding to the compact; and no state not belonging to the House of Bourbon was to be allowed to accede.

The cooperation between the French and Spanish Bourbons came to a miserable end during the French Revolutionary and Napoleonic Wars, and the later decades of the 19th century brought new complications. A French Bourbon prince led a force into Spain in 1823 to crush the liberalism to which Ferdinand VII was succumbing; but such Bourbon solidarity could not survive two events which were to rend both the Spanish and the French houses.

First, in March 1830, Ferdinand VII of Spain announced the revocation of the Salic Law of Succession, which Philip V had introduced into Spain in 1713. This meant that the sonless Ferdinand could be succeeded not by his brother Don Carlos conde de Molina but by his elder daughter Isabella (born after the revocation); and though Ferdinand temporarily reinstated the Salic Law in September 1832, he revoked it again 13 days later. On his death in 1833 the partisans of the disappointed Don Carlos started the first of the Carlist Wars in protest against Isabella's accession (see again SPAIN).

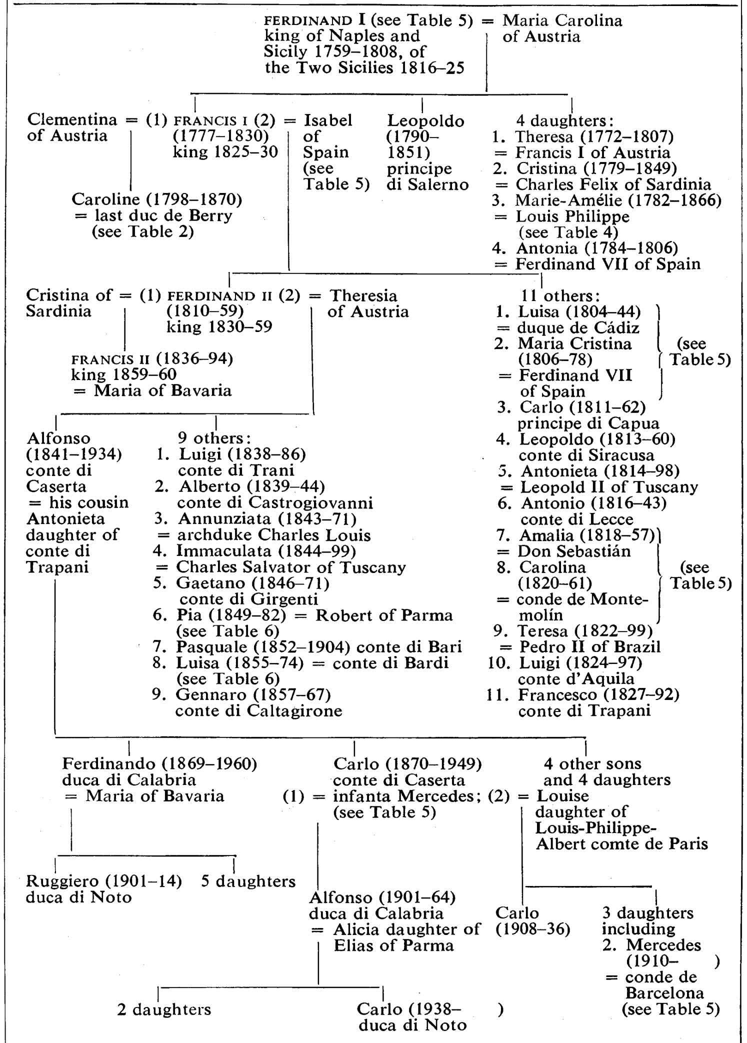
Carlism and the House of Orléans

Secondly, in France, the July Revolution of 1830 overthrew the "legitimate" Bourbon monarchy and transferred the throne to Louis-Philippe, head of the collateral line of Orléans. Odious enough already because Louis-Philippe's father, the self-styled Philippe Égalité, had voted in 1793 for the death sentence on Louis XVI, the House of Orléans became, by the usurpation of 1830, so much more odious to the Legitimists that some of the latter, when the "legitimate" male of France died out with the comte de Chambord in 1883, declined to recognize the head of the House of Orléans as the rightful pretender to France, as indeed he now was if the renunciation of 1713 was still to be observed (see again Tables 3 and 4); instead they preferred to disregard that renunciation and so to regard a Spanish prince as their rightful king. These Legitimists were known in France as "Blancs d'Espagne" ("Spanish Whites"). Most Legitimists, however, followed the final advice of the comte de Chambord by recognizing the rights of the House of Orléans to France.

While the dispossessed Bourbons—Spanish Carlists and French Legitimists—naturally sympathized with each other, their opponents—Queen Isabella and the House of Orléans—conversely gravitated together. One result was the crisis of the "Spanish Marriages" in the 1840s. While both Queen Isabella and her sister Luisa remained unmarried, the Spanish succession was an open prospect of great interest to governments concerned with maintaining the balance of power in Europe. If both sisters had married princes of the House of Orléans, as Louis-Philippe and the sister's mother, Maria Cristina, had originally suggested, French influence over Spain would have become too strong for the liking of the British government, which proposed instead that Isabella should marry Prince Leopold of Saxe-Coburg (more intimately linked with Great Britain than with France). Then, in 1843, the French and the British came to an understanding: Isabella should marry some "neutral" prince, preferably a Spanish Bourbon cousin; and only after the birth of a child to Isabella should Luisa marry Louis-Philippe's son Antoine duc de Montpensier. Of Isabella's eligible cousins, the conte de Montemolín was disfavoured by the Spanish government as a Carlist; the next senior was the doubtfully virile Don Francisco de Asis, who was generally thought unlikely to become a father; the third was Don Enrique duque de Sevilla, whose outspoken liberalism recommended him to the British government but not to the Spanish. Inadvertently, however, the British government in 1846 gave the French the impression that it was still secretly trying to press Prince Leopold on Spain, and the French reacted by arranging the Spanish marriages in a way quite contrary to British desire: Isabella and Luisa were married on the same day, October 10, 1846, to Don Francisco de Asis and to Montpensier, respectively. The immediate upshot was that the House of Orléans, apparently intending that Montpensier or a son of his should eventually be king of Spain, incurred the serious resentment of its former friends in Great Britain.

Isabella, who would have preferred to marry Don Enrique, spent conspicuously long periods apart from her consort and behaved indiscreetly with other men. When she bore a son in 1857, ill-wishers had little difficulty in casting doubts on his paternity. These doubts served the purposes of the extreme Carlists when the male line of Don Carlos died out in 1936, because they could argue that Isabella's male descendants were not those of Don Francisco de Asis—whose issue, under Salic Law,

Table 7: The Bourbons of the Two Sicilies



FERDINAND I (see Table 5) = Maria Carolina
king of Naples and of Austria
Sicily 1759–1808, of
the Two Sicilies 1816–25

Clementina = (1) FRANCIS I (2) = Isabel
of Austria (1777–1830) of
king 1825–30 Spain
(see
Table 5)

Caroline (1798–1870)
= last duc de Berry
(see Table 2)

Leopoldo
(1790–
1851)
principe
di Salerno

4 daughters:
1. Theresa (1772–1807)
= Francis I of Austria
2. Cristina (1779–1849)
= Charles Felix of Sardinia
3. Marie-Amélie (1782–1866)
= Louis Philippe
(see Table 4)
4. Antonia (1784–1806)
= Ferdinand VII of Spain

Cristina of = (1) FERDINAND II (2) = Theresia
Sardinia (1810–59) of Austria
king 1830–59

FRANCIS II (1836–94)
king 1859–60
= Maria of Bavaria

11 others:
1. Luisa (1804–44)
= duque de Cádiz
2. Maria Cristina
(1806–78)
= Ferdinand VII
of Spain
(see Table 5)
3. Carlo (1811–62)
principe di Capua
4. Leopoldo (1813–60)
conte di Siracusa
5. Antonieta (1814–98)
= Leopold II of Tuscany
6. Antonio (1816–43)
conte di Lecce
7. Amalia (1818–57)
= Don Sebastián
8. Carolina
(1820–61)
= conde de Montemolín
(see Table 5)
9. Teresa (1822–99)
= Pedro II of Brazil
10. Luigi (1824–97)
conte d'Aquila
11. Francesco (1827–92)
conte di Trapani

Alfonso
(1841–1934)
conte di
Caserta
= his cousin
Antonieta
daughter of
conte di
Trapani

9 others:
1. Luigi (1838–86)
conte di Trani
2. Alberto (1839–44)
conte di Castrogiovanni
3. Annunziata (1843–71)
= archduke Charles Louis
4. Immaculata (1844–99)
= Charles Salvator of Tuscany
5. Gaetano (1846–71)
conte di Girgenti
6. Pia (1849–82) = Robert of Parma
(see Table 6)
7. Pasquale (1852–1904) conte di Bari
8. Luisa (1855–74) = conte di Bardi
(see Table 6)
9. Gennaro (1857–67)
conte di Caltagirone

Ferdinando (1869–1960)
duca di Calabria
= Maria of Bavaria

Carlo (1870–1949)
conte di Caserta
(1) = infanta Mercedes; (2) = Louise
(see Table 5)

4 other sons
and 4 daughters

Louise
daughter of
Louis-Philippe-Albert comte de Paris

Ruggiero (1901–14) 5 daughters
duca di Noto

Alfonso (1901–64)
duca di Calabria
= Alicia daughter of
Elias of Parma

Carlo
(1908–36)

3 daughters
including
2. Mercedes
(1910–)
= conde de
Barcelona
(see Table 5)

2 daughters

Carlo (1938–)
duca di Noto

Carlism and Bourbon-Parma

would have been the next male heirs. Nearly all the other Bourbon princes, however, either had already recognized Isabella's rights or were maintaining incompatible pretensions to other thrones. The Carlists therefore had to look far afield in their search for a new pretender.

Certain princes of Bourbon-Parma responded to Carlist overtures but did not at the same time renounce their Parmesan titles, which under the settlement of 1748 were incompatible with claims to sovereignty over Spain. Thus they incurred the displeasure of the House of Orléans, which had to respect the settlement of 1748 because its own pretension to France depended on the analogous settlement of 1713.

BIBLIOGRAPHY. An introduction is provided by SANCHE DE GRAMONT, *Epitaph for Kings* (1968). Those interested in genealogy many readily consult HENRI VRIGNAULT, *Généalogie de la maison de Bourbon,* 2nd ed. rev. (1957), with 12 tables, and, on the bastard branches, *Légitimés de France de la maison de Bourbon de 1594 à 1820* (1965). Vrignault's work of 1957 may be brought up to date through the successive volumes of the periodical *Genealogisches Handbuch des Adels,* series "Fürstliche Häuser," whose first article "Bourbon" appeared in 1953 (vol. 2 of the series, 3 of the collection). While the story of the French Bourbon sovereigns is too vast to be resumed collectively in any single work other than a history of France, there are helpful works in English on the Spanish Bourbons and on the Neapolitan: SIR CHARLES PETRIE, *The Spanish Royal House* (1958); supplemented by THEO ARONSON on the Carlist question, *Royal Vendetta: The Crown of Spain, 1829–1965* (1966); and HAROLD ACTON, *The Bourbons of Naples, 1734–1825* (1956, reprinted 1974), and *The Last Bourbons of Naples, 1825–1861* (1961). JOHN D. BERGAMINI, *The Spanish Bourbons* (1974); and DESMOND SEWARD, *The Bourbon Kings of France* (1976), are works for the general reader. On the physiological aspect, the reader may first consult ANDRÉ DE MARICOURT and MAURICE DE BERTRANDFOSSE, *Les Bourbons, 1518–1830: hérédités, pathologie, amours et grandeur* (1936).

(J.R.-S./Ed.)

Brazil

The Federative Republic of Brazil (Portuguese: República Federativa do Brasil), the fifth largest country in the world, is exceeded in size only by the Soviet Union, Canada, China, and the United States. It covers an area of 3,286,487 square miles (8,511,965 square kilometres), or about half of all South America. Brazil is so vast that, although it faces the Atlantic Ocean along 4,600 miles (7,400 kilometres) of coastline, it also borders on every nation of the South American continent except Chile and Ecuador. Stretching for approximately 2,700 miles from north to south and from east to west, it contains no desert, high mountain, or Arctic environments that limit the extent of human occupancy.

Most of the inhabitants of Brazil, the world's sixth most populous nation, are concentrated along the eastern seaboard. Its national capital, Brasília, however, is located inland, on the outer edge of intensive settlement. Planned from scratch, Brasília replaced Rio de Janeiro as the capital in 1960. Unique among the nations of the Americas, this former colony of Portugal did not become fragmented into separate countries, as did the British and Spanish possessions, but retained its identity through the intervening centuries and a variety of forms of government. The Portuguese language is, therefore, universal except among a few thousand native Indians in the most remote reaches of the Amazon River system.

The physical and human features of this broad land are rich and various. Because the greater part of the Amazon Basin lies within Brazilian territory, scientists from throughout the world are attracted to the region to study the Earth's largest river system and most extensive virgin rain forest. Brazil has enormous expanses of untouched tropical forest and sparsely settled savannas with a scattering of isolated villages. It is a land rich in national resources whose burgeoning cities, huge hydroelectric and industrial complexes, mines, and fertile farmlands have placed it among the world's most productive countries. In contrast, however, Brazil is also a country that has had to struggle with a rapidly growing population, an unstable economy, soaring inflation, and a highly volatile political life.

This article is divided into the following sections:

Physical and human geography

THE LAND

Relief. The country can be readily described in terms of its highlands and their adjoining lowlands, which are separated from each other by an almost continuous series of physical barriers, the most impressive being the Great Escarpment.

The Great Escarpment. The massive fault block called the Great Escarpment extends for 1,600 miles along the eastern coast. From Rio de Janeiro southward it is known as the Serra do Mar; this segment of the Great Escarpment, presenting an almost sheer face to the sea, averages about 3,000 feet (1,000 metres) in elevation. It is flanked on the east by the tips of lower fault blocks, including Rio de Janeiro's Sugar Loaf (Pão de Açúcar) and Gávea mountains, and a string of islands, including the large ones of Santa Catarina, São Sebastião, and Grande.

Serra do Mar

Northward from Rio de Janeiro the escarpment breaks into a series of highly faulted and fractured blocks characterized by deep, steep-sided valleys ringed by high, rounded peaks and nearly vertical slabs of bare rock. It is known variously as the Orgãos Mountains northwest of Rio de Janeiro, the Aimorés Mountains along the Minas Gerais-Espírito Santo border, and, in Bahia, as the Geral Mountains. The barrier finally ends in a low scarp. Elsewhere, the escarpment ends in a rim marked by fall lines, where the rivers, forming rapids and waterfalls, plunge onto the coastal lowlands.

Lowlands. The Brazilian lowlands are composed largely of three principal regions: the Amazon lowlands, the west-central Pantanal, and the coastal lowlands. The largest of these regions by far is the Amazon lowlands, the greater proportion of which consists of the hilly *terra firme,* land composed of alluvium that was deposited as much as 2,500,000 years ago and uplifted to its present position above flood level. The most striking lowland features, however, are the annually flooded plains, or *várzeas,* along the Amazon and the Paraguay river systems. The widely dispersed grasslands of the Amazon lowlands exhibit conditions of monotonously low relief, but the active floodplains that border the Amazon River and its tributaries comprise a complex of features; these range from shallow lakes and swamps occupying old watercourses, through

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Goiás	12 00 s 48 00 w
Maranhão	05 00 s 45 00 w
Mato Grosso	13 00 s 56 00 w
Mato Grosso do Sul	21 00 s 55 00 w
Minas Gerais	18 00 s 44 00 w
Pará	04 00 s 53 00 w
Paraíba	07 15 s 36 30 w
Paraná	24 00 s 51 00 w
Pernambuco	08 00 s 37 00 w
Piauí	07 00 s 43 00 w
Rio de Janeiro	22 00 s 42 30 w
Rio Grande do Norte	05 45 s 36 00 w
Rio Grande do Sul	30 00 s 54 00 w
Rondônia	11 00 s 63 00 w
Roraima	01 00 n 61 00 w
Santa Catarina	27 00 s 50 00 w
São Paulo	22 00 s 49 00 w
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Abaetetuba	01 42 s 48 54 w
Açu	05 34 s 36 54 w
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Bagé	31 20 s 54 06 w
Balsas	07 31 s 46 02 w
Barbacena	21 14 s 43 46 w
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Barra do Garças	15 53 s 52 15 w
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Barretos	20 33 s 48 33 w
Bauru	22 19 s 49 04 w
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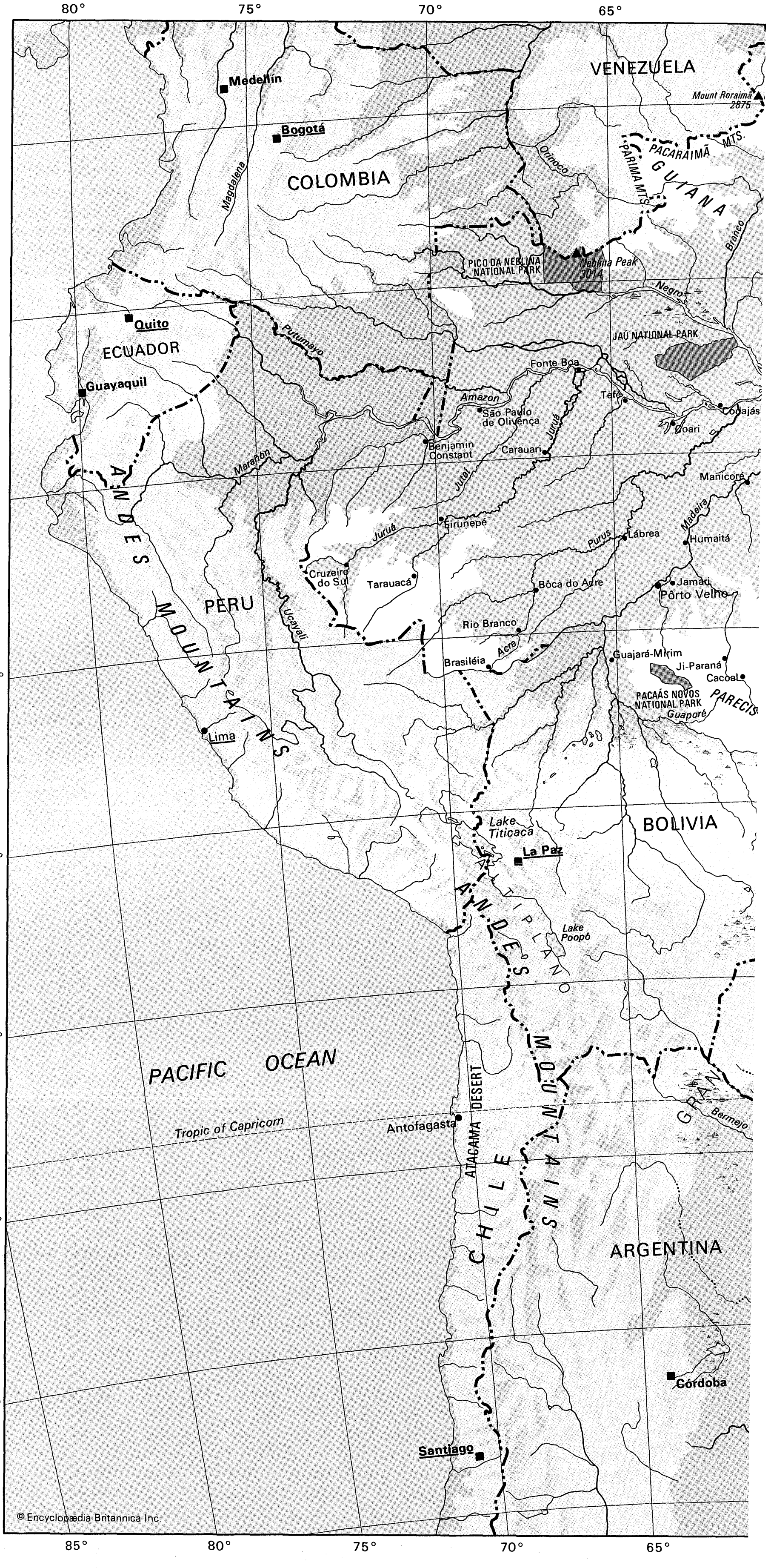
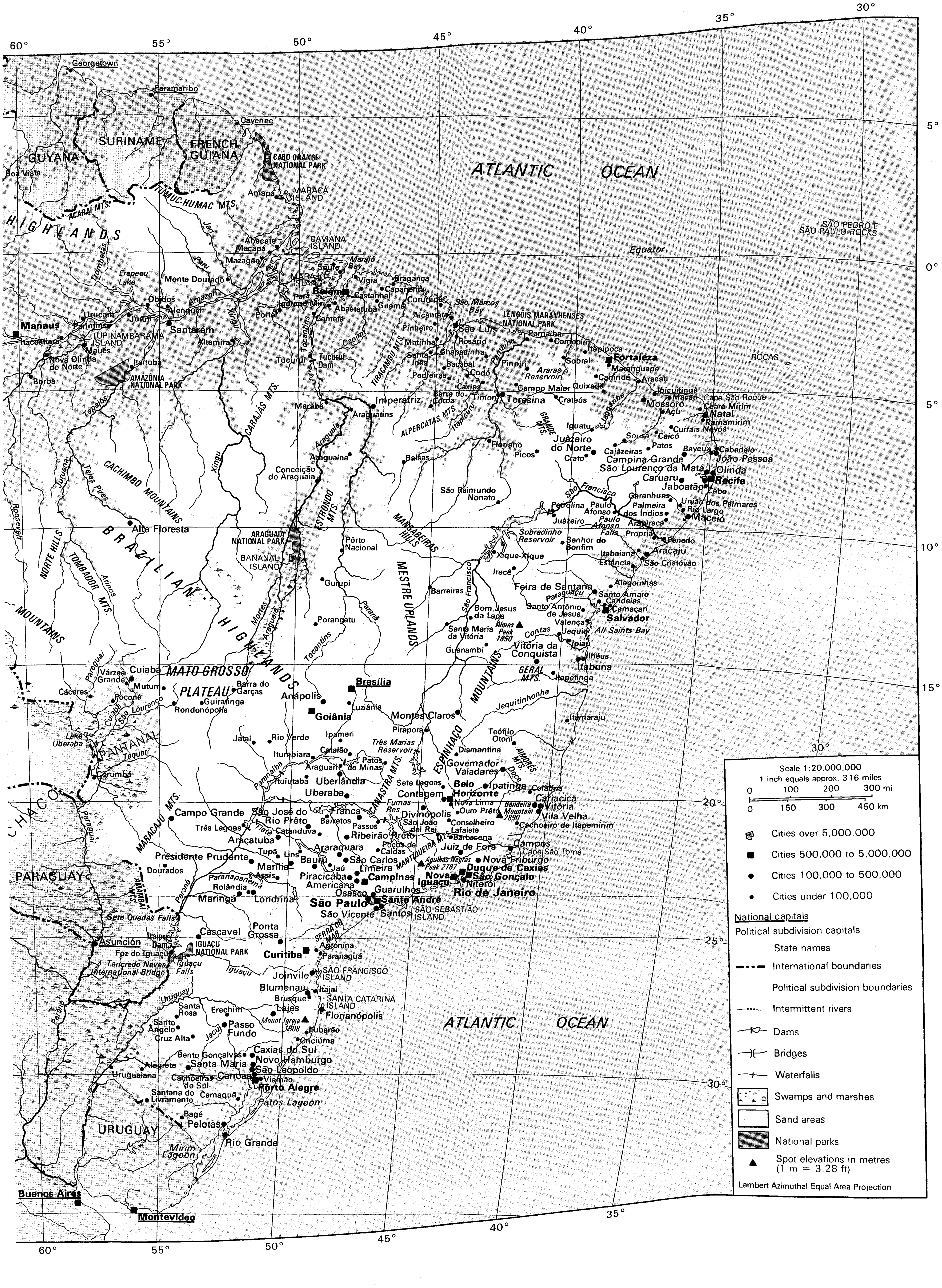
Benjamin Constant	04 22 s 70 02 w
Bento Gonçalves	29 10 s 51 31 w
Blumenau	26 56 s 49 03 w
Boa Vista	02 49 n 60 30 w
Bôca do Acre	08 45 s 67 23 w
Bom Jesus da Lapa	13 15 s 43 25 w
Borba	04 24 s 59 35 w
Bragança	01 03 s 46 46 w
Brasiléia	11 00 s 68 44 w
Brasília	15 47 s 47 55 w
Brusque	27 06 s 48 56 w
Cabedelo	06 58 s 34 50 w
Cabo	08 17 s 35 02 w
Cáceres	16 04 s 57 41 w
Cachoeira do Sul	30 02 s 52 54 w
Cachoeiro de Itapemirim	20 51 s 41 06 w
Cacoal	11 30 s 61 25 w
Caicó	06 27 s 37 06 w
Cajàzeiras	06 54 s 38 34 w
Camaçari	12 41 s 38 18 w




ATLANTIC OCEAN
Equator
SÃO PEDRO E SÃO PAULO ROCKS
ROCAS
Georgetown
Paramaribo
Cayenne
GUYANA
SURINAME
FRENCH GUIANA
CABO ORANGE NATIONAL PARK
Boa Vista
ACARAÍ MTS.
TUMUC-HUMAC MTS.
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CAVIANA ISLAND
Abacate
Macapá
Mazagão
Marajó Bay
Soure
MARAJÓ ISLAND
Vigia
Bragança
Capanema
Belém
Castanhal
Pará
Igarapé-Miri
Abaetetuba
Guamá
Portel
Cametá
Monte Dourado
Erepecu Lake
Óbidos
Alenquer
Amazon
Urucará
Juruti
Manaus
Parintins
Itacoatiara
Santarém
TUPINAMBARAMA ISLAND
Maués
Nova Olinda do Norte
Borba
Itaituba
AMAZÔNIA NATIONAL PARK
Altamira
Tucuruí
Tucuruí Dam
Trombetas
Jari
Paru
Xingu
Tocantins
Capim
Tapajós
Juruena
Teles Pires
Roosevelt
CARAJÁS MTS.
TIRACAMBU MTS.
Curutupu
Alcântara
São Marcos Bay
LENÇÓIS MARANHENSES NATIONAL PARK
Pinheiro
São Luís
Matinha
Rosário
Santa Inês
Chapadinha
Bacabal
Pedreiras
Codó
Caxias
Parnaíba
Camocim
Itapipoca
Fortaleza
Maranguape
Canindé
Aracati
Sobral
Piripiri
Araras Reservoir
Campo Maior
Quixadá
Teresina
Timon
Barra do Corda
Imperatriz
Marabá
Araguatins
ALPERCATAS MTS.
Itapicuru
Crateús
Jaguaribe
Icapuí
Mossoró
Açu
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Vitória da Conquista
Ipiaú
Ilhéus
Itabuna
GERAL MTS.
Itapetinga
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São Paulo
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São Vicente
Santos
SÃO SEBASTIÃO ISLAND
Sete Quedas Falls
Ponta Grossa
Cascavel
SERRA DO MAR
Itaipu Dam
Asunción
Foz do Iguaçu
IGUAÇU NATIONAL PARK
Curitiba
Antonina
Paranaguá
Tancredo Neves International Bridge
Iguaçu Falls
Iguaçu
Joinvile
SÃO FRANCISCO ISLAND
Blumenau
Itajaí
Brusque
Uruguay
Santa Rosa
Erechim
Lajes
SANTA CATARINA ISLAND
Florianópolis
Mount Igreja 1808
Santo Ângelo
Passo Fundo
Tubarão
Jacuí
Cruz Alta
Criciúma
Caxias do Sul
Bento Gonçalves
Novo Hamburgo
Santa Maria
São Leopoldo
Alegrete
Uruguaiana
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Canoas
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Camaquã
Patos Lagoon
Bagé
Pelotas
URUGUAY
Mirim Lagoon
Rio Grande
Buenos Aires
Montevideo
60°
55°
50°
45°
40°
35°
30°
5°
0°
5°
10°
15°
20°
25°
30°
Scale 1:20,000,000
1 inch equals approx. 316 miles
0 100 200 300 mi
0 150 300 450 km
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Cities 500,000 to 5,000,000
Cities 100,000 to 500,000
Cities under 100,000
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Colatina	19 32 s 40 37 w
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Contagem	19 55 s 44 06 w
Corumbá	19 01 s 57 39 w
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Gurupi	11 43 s 49 04 w
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Itabuna	14 48 s 39 16 w
Itacoatiara	03 08 s 58 25 w
Itaituba	04 17 s 55 59 w
Itajaí	26 53 s 48 39 w
Itamaraju	17 04 s 39 32 w
Itapetinga	15 15 s 40 15 w
Itapipoca	03 30 s 39 35 w
Ituiutaba	18 58 s 49 28 w
Itumbiara	18 25 s 49 13 w
Jaboatão	08 07 s 35 01 w
Jamari	08 45 s 63 27 w
Jataí	17 53 s 51 43 w
Jaú	22 18 s 48 33 w
Jequié	13 51 s 40 05 w
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Macau	05 07 s 36 38 w
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Marabá	05 21 s 49 07 w
Maranguape	03 53 s 38 40 w
Marília	22 13 s 49 56 w
Maringá	23 25 s 51 55 w
Matinha	03 06 s 45 02 w
Maués	03 24 s 57 42 w
Mazagão	00 07 s 51 17 w
Monte Dourado	00 52 s 52 31 w
Montes Claros	16 43 s 43 52 w
Mossoró	05 11 s 37 20 w
Mutum	15 48 s 54 53 w
Natal	05 47 s 35 13 w
Niterói	22 53 s 43 07 w
Nova Friburgo	22 16 s 42 32 w
Nova Iguaçu	22 45 s 43 27 w
Nova Lima	19 59 s 43 51 w
Nova Olinda do Norte	03 45 s 59 03 w
Novo Hamburgo	29 41 s 51 08 w
Óbidos	01 55 s 55 31 w
Olinda	08 01 s 34 51 w
Osasco	23 32 s 46 46 w
Ouro Prêto	20 23 s 43 30 w
Palmeira dos Índios	09 25 s 36 37 w
Pará, see Belém	
Paranaguá	25 31 s 48 30 w
Parintins	02 36 s 56 44 w
Parnaíba	02 54 s 41 47 w
Parnamirim	05 55 s 35 15 w
Passo Fundo	28 15 s 52 24 w
Passos	20 43 s 46 37 w
Patos	07 01 s 37 16 w
Patos de Minas	18 35 s 46 32 w
Paulo Afonso	09 21 s 38 14 w
Pedreiras	04 34 s 44 39 w
Pelotas	31 46 s 52 20 w
Penedo	10 17 s 36 36 w
Petrolina	09 24 s 40 30 w
Picos	07 05 s 41 28 w
Pinheiro	02 31 s 45 05 w
Piracicaba	22 43 s 47 38 w
Pirapora	17 21 s 44 56 w
Piripiri	04 16 s 41 47 w
Poconé	16 15 s 56 37 w
Poços de Caldas	21 48 s 46 34 w
Ponta Grossa	25 05 s 50 09 w
Porangatu	13 26 s 49 10 w
Portel	01 57 s 50 49 w
Pôrto Alegre	30 04 s 51 11 w
Pôrto Nacional	10 42 s 48 25 w
Pôrto Velho	08 46 s 63 54 w
Presidente Prudente	22 07 s 51 22 w
Propriá	10 13 s 36 51 w
Quixadá	04 58 s 39 01 w
Recife	08 03 s 34 54 w
Ribeirão Prêto	21 10 s 47 48 w
Rio Branco	09 58 s 67 48 w
Rio de Janeiro	22 54 s 43 14 w
Rio Grande	32 02 s 52 05 w
Rio Largo	09 29 s 35 51 w
Rio Verde	17 43 s 50 56 w
Rolândia	23 18 s 51 22 w
Rondonópolis	16 28 s 54 38 w
Rosário	02 57 s 44 14 w
Salvador	12 59 s 38 31 w
Santa Inês	03 39 s 45 22 w
Santa Maria	29 41 s 53 48 w
Santa Maria da Vitória	13 24 s 44 12 w
Santa Rosa	27 52 s 54 29 w
Santana do Livramento	30 53 s 55 31 w
Santarém	02 26 s 54 42 w
Santo Amaro	12 32 s 38 43 w
Santo André	23 40 s 46 31 w
Santo Ângelo	28 18 s 54 16 w
Santo Antônio de Jesus	12 58 s 39 16 w
Santos	23 57 s 46 20 w
São Carlos	22 01 s 47 54 w
São Cristóvão	11 01 s 37 12 w
São Gonçalo	22 51 s 43 04 w
São João del Rei	21 09 s 44 16 w
São José do Rio Prêto	20 48 s 49 23 w
São Leopoldo	29 46 s 51 09 w
São Lourenço da Mata	08 00 s 35 03 w
São Luís	02 31 s 44 16 w
São Paulo	23 32 s 46 37 w
São Paulo de Olivença	03 27 s 68 48 w
São Raimundo Nonato	09 01 s 42 42 w
São Vicente	23 58 s 46 23 w
Senhor do Bonfim	10 27 s 40 11 w
Sete Lagoas	19 27 s 44 14 w
Sobral	03 42 s 40 21 w
Soure	00 44 s 48 31 w
Sousa	06 45 s 38 14 w
Tarauacá	08 10 s 70 46 w
Tefé	03 22 s 64 42 w
Teófilo Otoni	17 51 s 41 30 w
Teresina	05 05 s 42 49 w
Timon	05 06 s 42 49 w
Três Lagoas	20 48 s 51 43 w
Tubarão	28 30 s 49 01 w
Tucuruí	03 42 s 49 27 w
Tupã	21 56 s 50 30 w
Uberaba	19 45 s 47 55 w
Uberlândia	18 56 s 48 18 w
União dos Palmares	09 10 s 36 02 w
Urucará	02 32 s 57 45 w
Uruguaiana	29 45 s 57 05 w
Valença	13 22 s 39 05 w
Várzea Grande	15 39 s 56 08 w
Viamão	30 05 s 51 02 w
Vigia	00 48 s 48 08 w
Vila Velha (Espírito Santo)	20 20 s 40 17 w
Vitória	20 19 s 40 21 w
Vitória da Conquista	14 51 s 40 51 w
Xique-Xique	10 50 s 42 44 w

Physical features and points of interest

Acaraí Mountains	01 50 N 57 40 w
Acre, *river*	08 45 s 67 22 w
Aimorés Mountains	19 00 s 41 00 w
Agulhas Negras Peak	22 23 s 44 38 w
All Saints Bay (Baía de Todos os Santos)	12 48 s 38 38 w
Almas Peak	13 33 s 41 56 w
Alpercatas Mountains	06 00 s 45 00 w
Amambaí Mountains	23 10 s 55 30 w
Amazon (Solimões), *river*	00 10 s 49 00 w
Amazõnia National Park	04 30 s 56 30 w
Araguaia, *river*	05 21 s 48 41 w
Araguaia National Park	10 30 s 50 15 w
Araras Reservoir	04 20 s 40 28 w
Arinos, *river*	10 25 s 58 20 w
Bananal Island	11 30 s 50 15 w
Bandeira Mountain	20 26 s 41 47 w
Branco, *river*	01 24 s 61 51 w
Brazilian Highlands	14 00 s 52 00 w
Cabo Orange National Park	03 00 N 51 15 w
Cachimbo Mountains	08 30 s 55 30 w
Canastra Mountains	20 00 s 46 20 w
Capim, *river*	01 40 s 47 47 w
Carajás Mountains	06 00 s 51 20 w
Caviana Island	00 10 N 50 10 w
Contas, *river*	14 17 s 39 01 w
Cuiabá, *river*	17 05 s 56 36 w
Doce, *river*	19 37 s 39 49 w
Erepecu Lake	01 20 s 56 35 w
Espinhaço Mountains	17 30 s 43 30 w
Estrondo Mountains	09 00 s 48 45 w
Furnas Reservoir	20 20 s 46 00 w
Geral Mountains	14 00 s 41 00 w
Grande Mountains	06 00 s 40 52 w
Guaíra Falls, see Sete Quedas Falls	
Guaporé, *river*	11 54 s 65 01 w
Guiana Highlands	04 00 N 60 00 w
Igreja, Mount	28 08 s 49 30 w
Iguaçu, *river*	25 36 s 54 36 w
Iguaçu Falls	25 41 s 54 26 w
Iguaçu National Park	25 25 s 54 00 w
Itaipu Dam	25 22 s 54 34 w
Itapicuru, *river*	07 28 s 47 30 w
Jacuí, *river*	30 02 s 51 15 w
Jaguaribe, *river*	04 25 s 37 44 w
Jari, *river*	01 09 s 51 54 w
Jaú National Park	02 15 s 62 40 w
Jequitinhonha, *river*	15 51 s 38 53 w
Juruá, *river*	02 37 s 65 44 w
Juruena, *river*	07 20 s 58 03 w
Jutaí, *river*	02 43 s 66 57 w
Lençóis Maranhenses National Park	02 30 s 43 00 w
Madeira, *river*	03 22 s 58 45 w
Mangabeiras Hills	10 00 s 46 30 w
Manso, see Mortes	
Mantiqueira Mountains	22 00 s 44 45 w
Mar, Serra do, *escarpment*	25 00 s 48 00 w
Maracá Island	02 05 N 50 25 w
Maracaju Mountains	21 00 s 55 01 w
Marajó Bay	01 00 s 48 30 w
Marajó Island	01 00 s 49 30 w
Mato Grosso Plateau	15 30 s 56 00 w
Mestre Uplands	12 30 s 46 10 w
Mirim Lagoon	32 45 s 52 50 w
Mortes (Manso), *river*	11 45 s 50 44 w
Neblina Peak	00 46 N 66 00 w
Negro, *river*	03 08 s 59 55 w
Norte Hills	11 20 s 59 00 w
Pacaás Novos National Park	11 15 s 63 30 w
Pacaraimã Mountains	04 05 N 61 30 w
Pantanal, *lowland*	18 00 s 56 00 w
Pará, *river*	01 30 s 48 55 w
Paraguaçu, *river*	12 45 s 38 54 w
Paraguai, *river*	27 18 s 58 38 w
Paraná, *river*	33 43 s 59 15 w
Paranã, *river*	12 30 s 48 14 w
Paranaíba, *river*	20 07 s 51 05 w
Paranapanema, *river*	22 40 s 53 09 w
Parecis Mountains	13 00 s 60 00 w
Parima Mountains	02 30 N 64 00 w
Parnaíba, *river*	03 00 s 41 50 w
Paru, *river*	01 33 s 52 38 w
Patos Lagoon	31 06 s 51 15 w
Paulo Afonso Falls	09 24 s 38 13 w
Pico da Neblina National Park	00 30 N 66 00 w
Purus, *river*	03 42 s 61 28 w
Rocas, *atoll*	03 52 s 33 49 w
Roosevelt, *river*	07 35 s 60 20 w

Roraimã, Mount	05 12 N	60 44 W
Santa Catarina Island	27 36 S	48 30 W
São Francisco, *river*	10 30 S	36 24 W
São Francisco Island	26 18 S	48 37 W
São Lourenço, *river*	15 52 S	56 04 W
São Manuel, see Teles Pires		
São Marcos Bay	02 36 S	44 28 W
São Pedro e São Paulo Rocks	00 56 N	29 22 W
São Roque, Cape	05 29 S	35 16 W
São Sebastião Island	23 50 S	45 18 W
São Tomé, Cape	21 59 S	40 59 W
Sete Quedas Falls (Guaíra Falls)	24 02 S	54 16 W
Sobradinho Reservoir	09 32 S	40 46 W
Solimões, see Amazon		
Tancredo Neves International Bridge	25 33 S	54 32 W
Tapajós, *river*	02 24 S	54 41 W
Taquari, *river*	19 15 S	57 17 W
Teles Pires (São Manuel), *river*	07 21 S	58 03 W
Tietê, *river*	20 40 S	51 35 W
Tiracambu Mountains	03 15 S	46 30 W
Tocantins, *river*	01 45 S	49 10 W
Todos os Santos, Baía de, see All Saints Bay		
Tombador Mountains	12 00 S	57 40 W
Três Marias Reservoir	18 12 S	45 15 W
Trombetas, *river*	01 55 S	55 35 W
Tucuruí Dam	03 48 S	49 25 W
Tumuc-Humac Mountains	02 20 N	55 00 W
Tupinambarama Island	03 00 S	58 00 W
Uberaba, Lake	17 30 S	57 45 W
Uruguay, *river*	34 12 S	58 18 W
Xingu, *river*	01 30 S	51 53 W

mangrove woods in the tidal flats of the delta, to newly deposited land built up on the inner sides of river bends and low but fertile and well-drained levees that may be inundated for only a few weeks each year.

The Pantanal

The region called the Pantanal, an extension of the Gran Chaco plain, is an area of swamps and marshes in northwestern Mato Grosso do Sul and southern Mato Grosso states; it is dissected by the upper Paraguay river system. Noted for its striking natural beauty and its rich wildlife, the Pantanal covers an area of some 39,000 square miles. During the rainy season the Paraguay overflows its banks, spreading great sheets of water that leave only the tops of scattered levees and low hills dry.

The Atlantic lowlands, which range up to 125 miles in width in the north and narrow and finally disappear to the south, provide a wide variety of relief features: level floodplains, levees, swamps, lagoons, sand dunes, and mile after mile of white, sandy beaches here and there protected by offshore coral reefs. Where the rocky slopes of the Great Escarpment plunge directly into the ocean, there are various deep harbours. Some of the most famous harbours are Guanabara Bay, on which Rio de Janeiro and Niterói are situated, and All Saints Bay, on which Salvador is located. The Amazon itself provides deep water for oceangoing ships as far inland as Manaus and for smaller ships as far as Iquitos in eastern Peru. Another lowland feature is the older uplifted tertiary floodplain that is known as the Amazon Plateau in the north and the Barreiras Formation along the coast; it has heavy clay beds that form a nearly level surface and intermittently abut the rim of the highlands.

The highlands. The highlands, which make up the bulk of Brazil's landmass, include the Brazilian, or Central, Highlands (which are also called Planalto Central) and the country's share of the Guiana Highlands in the far north. By far the largest in area is the Brazilian Highlands. During the Paleozoic Era these highlands were peneplained—that is, they were worn down to an ancient sea level and then uplifted to their present elevation. The topography of this surface varies from level-topped plateaus, through rolling hills, to deeply incised valleys, above which rise the dissected relics of once-mighty mountain ranges that are the source of most of Brazil's abundant and varied mineral wealth.

Although none of these relic ranges is particularly high in elevation, nowhere surpassing 10,000 feet, the local relief is often extreme and the land exceptionally rugged. The Mantiqueira Mountains in southern Minas Gerais reach 9,143 feet at the Agulhas Negras Peak on the Rio de Janeiro state border, and 9,482 feet at Bandeira Peak, one of Brazil's highest points, near the Aimorés Mountains. The Espinhaço Mountains extend from central Minas Gerais into southern Bahia, where Almas Peak reaches 6,070 feet. Along the Venezuelan border Mount Roraimã attains 9,432 feet in the Guiana Highlands, and Neblina Peak, discovered in 1962, reaches 9,888 feet (3,014 metres), the highest point in Brazil.

The erosion of these and other ranges at various times deposited thick beds of sedimentary rock along the outer margins of the central core. In the south these earlier formations have been buried by diabase lava flows, abutting against the rim of the Great Escarpment and extending southward as the Geral Mountains (not the same as those of Bahia). In the west the diabase ends in a steep rim over which the Iguaçu River tumbles to form the spectacular Iguaçu Falls just before converging with the Paraná River. The Guaíra Falls were a similar attraction until the huge hydroelectric dam at Itaipú was built downstream on the Paraná River, causing the falls to be submerged.

Major river systems

Drainage. The Brazilian Highlands are drained by three major river systems, the great Amazon system in the north, the Paraguay–Paraná–Plata system in the south, and the São Francisco, which flows to the northeast. The major systems all drain into the Atlantic Ocean, as do a number of smaller rivers in the east.

The Amazon River rises from a point in the Peruvian Andes, within 100 miles of the Pacific Ocean, whence its course meanders about 4,000 miles to the Atlantic, into which it spills some 20 percent of all the Earth's fresh water. Its succession of great tributaries includes the Tocantins–Araguaia, the Xingu, the Tapajós, and the Madeira on the southern side, and the Negro River on the northern side. The Amazon lowland has the peculiarity of being widest along the eastern base of the Andes. It becomes narrower toward the east, until, east of Manaus, only a narrow ribbon of floodplain separates the Guiana Highlands to the north from the Brazilian Highlands to the south. The floodplain opens up again as it approaches the Atlantic. The Amazon is navigable all the way into eastern Peru, but the tributaries are all interrupted by falls and rapids where they descend from the highlands. In contrast to the Paraguay–Paraná–Plata system, that of the Amazon River has not been harnessed for power production except at the Tucuruí Dam, which was built in the mid-1980s across the Tocantins River south of Belém and near the Carajás Grande mining complex.

The second of the great river systems of Brazil is that of the Paraguay–Paraná–Plata. From southwestern Minas Gerais southward, the highlands are drained into the Paraná, eventually reaching the sea through the Río de la Plata. Brazil's two southernmost states are drained through the Uruguay River also to the Plata. These rivers in Brazil are navigable only for short stretches.

São Francisco River

The São Francisco River, the largest river wholly within Brazil, is the third of the country's great river systems. The São Francisco rises in the Brazilian Highlands in western

Union Press/Bruce Coleman, Inc., New York City

Canoe on the Negro River in the Amazon rain forest, Amazonas state, northern Brazil.

Minas Gerais and southern Goiás, inland from Rio de Janeiro. It flows for more than 1,000 miles northward into the state of Bahia and the border of Pernambuco before it turns eastward to the Atlantic Ocean. The river is navigable for shallow-draft riverboats between Pirapora in Minas Gerais north of Belo Horizonte and Juàzeiro in Bahia. Below Juàzeiro the river current becomes swifter as it approaches the Paulo Afonso Falls. Only the lower river is navigable for ocean ships.

Numerous rivers drain directly into the Atlantic eastward from the Brazilian interior, but, although they pass through the most populated part of the country, most of them are too short, steep, or unreliable even for hydroelectric developments, much less for water-borne traffic. The Paranaíba in the state of Piauí and the Jacuí in Rio Grande do Sul are the most important of these for navigability.

Climate. Brazil has a humid tropical and subtropical climate except for a semiarid area in the northeast, sometimes called the drought quadrilateral or drought polygon, that extends from northern Bahia to the coast between Natal and São Luís and receives only between 15 and 30 inches (375 and 750 millimetres) of precipitation a year. Most of the rest of the country gets a moderate amount of rainfall, between 40 and 70 inches a year; exceptions to this are certain parts of the Amazon Basin and the sea-facing rim of the Serra do Mar, where the rains are much heavier.

Average temperatures

Temperatures in Brazil are remarkably uniform during the summer months (November to April), averaging about 79° F (26° C) over most of the lowlands in January, and a few degrees less in the highlands, depending upon elevation. The coast of Rio Grande do Sul is also somewhat cooler, averaging around 73° F (23° C), whereas the interior of the drought quadrilateral, the hottest region of the country, averages about 84° F (29° C), with daytime temperatures over 100° F (38° C) in summer. Winter temperatures (May to October) in the Amazon lowland remain virtually unchanged from those of the summer months, but they drop to about 79° F (about 26° C) in the drought quadrilateral. In the Brazilian Highlands winter temperatures are about 68° F (20° C) and are even lower toward the south where Curitiba, at 3,000-feet elevation, averages 57° F (14° C) in June and July. The mean for those months at Pôrto Alegre is the same, but, with greater distance from Antarctic air masses in the lowlands to the north, temperatures rise rapidly, averaging 73° F (23° C) at Rio de Janeiro.

Summer air masses. In typical monsoonal fashion, the high-pressure system over the Brazilian Highlands during the summer months is replaced by a low-pressure one that pulls in warm moist air from both the South Atlantic and the North Atlantic high-pressure systems, the latter picking up more moisture as it passes over the forests of the Amazon region. Where these meet along the intertropical convergence zone (ITCZ), torrential downpours occur, amounting to more than 12 inches a month; thus, central Brazil has a wet-dry tropical climate. These are critical months in the semiarid northeastern region, which normally receives its precipitation from the easternmost edge of the shifting ITCZ; in more than half of the years, however, it either swings too far east, causing floods, or does not reach the area at all, bringing drought that is sometimes prolonged for years. This condition brings much misery to the sertão, the northeastern backcountry, and has become a major cause for migration from the sertão to the cities.

Winter air masses. During the winter months southernmost Brazil is dominated by polar air masses and their associated cyclonic storms. These generate four to five inches of monthly precipitation, occasionally in the form of snow, accompanied by regular frosts in the south and bringing near-freezing temperatures as far north as São Paulo. Further penetration is, however, blocked by a high-pressure system that forms over the Brazilian Highlands in winter, keeping precipitation there down to a desertlike measure of about one-half inch a month. Modified polar air, skirting this system, may bring cool, rainy weather along the coast as far north as Recife and, in the west, to the Pantanal. Cool air occasionally spills over from the Paraguay lowland into that of the Amazon as far north as the Guyana border, dropping temperatures by as much as 9° F (5° C).

Soils. Although complex in detail, Brazil's soils can generally be categorized as fertile and infertile.

Infertile soils. Because of the great age of so many of the country's landforms and its tropical climate, a high proportion of its soils have been intensely leached, leaving them poor in nutrients and organic matter. This is most evident on the surfaces of the Brazilian Highlands and Tertiary Plateaus, where soils, made up almost entirely of iron and aluminum silicates, can be as much as 90 feet deep, with laterites marking present or earlier water tables. Relic laterite layers on the surface may impede cultivation.

Amazonian soils

Soils of the Amazonian *terra firme* and similar landforms in the country are also leached but usually not as deeply. Their fertility depends on the closed ecological system of the rain forest, in which high temperatures and precipitation cause the immediate decay and recycling of dead organic matter. Once this cycle is broken, there is an almost immediate loss of nutrients and organic matter, and the soils quickly become sterile. For this reason, much of the once luxuriant coastal forest area is now an impoverished grassland, and there is fear among ecologists that the same fate awaits the Amazonian rain forest if its clearing continues.

Fertile soils. Within these broad zones of leached soils there are more fertile ones that have, for the most part, been created by erosion and the deposition of fresh silt along Brazil's numerous watercourses; they are scattered, however, and often not very extensive. In the highlands, fertile soils appear under two circumstances: where the ancient soils have been stripped away by erosion and younger, shallower soils have formed over the underlying bedrock, or where nutrients and organic matter have accumulated in the bottoms of swales and shallow valleys. These lower lands are usually covered with gallery forests because of the year-round availability of water, and their decaying litter adds to the fertility of the soil.

There are two major soil types of high fertility that are continuous over large areas. In the dry northeastern interior, leaching has been retarded and many of the original nutrients remain; however, these soils require more water than is generally available to make them agriculturally productive. By far the most fertile soils are located over the diabase formations that extend northward from central Rio Grande do Sul into southern Minas Gerais. There the rocks weather into a deep reddish-purple soil, known as *terra roxa,* that is highly productive even under humid conditions and is especially noted for the rich Brazilian coffee it produces.

Plant and animal life. Except in the more remote and rugged sections of Brazil's eastern highlands, little of the original flora and fauna remain south and east of the scrublands of the Brazilian Highlands and the semiarid northeastern region. The once luxuriant hardwood forests that dominated the eastern seaboard and the valleys of the interior have long since been destroyed, as have the magnificent Paraná pine, or *Araucaria,* forests that once covered the diabase plateau as far north as the terminus of regular winter frost. Monkeys, parrots, and other exotic wildlife are found only in zoos, private menageries, or small patches of relic habitats, where the original flora and fauna remain. The coastal waterways and swamps that once teemed with wildfowl and alligators have been replaced by salt pans, marinas, and condominiums.

Grassland and scrub. The grasslands of Rio Grande do Sul have been largely replaced by artificial pastures and wheat fields, but those of the Pantanal remain as natural grazing lands. There the many sloughs and watercourses still support an abundance of wildlife, including the giant pirarucu, a fish that is herded into enclosures like underwater cattle pens until needed for food. Wildfowl also abound in the Pantanal and include the ibis and the heron and wild ducks and migratory geese. On higher ground, and continuing on through the savannas of central Brazil, are the ubiquitous rheas, South American ostriches, and *siriemas,* a type of roadrunner, as well as a variety of

Pantanal fauna

game birds such as quail and partridge. There are numerous species of snakes, including the deadly fer-de-lance (jararaca) and the rattlesnake, a wide variety of lizards, some as much as a yard long, armadillos, and anteaters that also prey on termite nests, which may stand more than six feet high.

The vegetation of the savanna lands of Brazil varies considerably from treeless plains of coarse bunchgrasses and low woods of gnarled and stunted trees, where there is limited water, to gallery forests, where there is a year-round water supply. These savannas have no herds of game animals like their African counterparts, although in times past deer were plentiful. Until the introduction of cattle, deer were the principal prey of the *onças,* a type of jaguar, and of the ocelots that inhabited the edges of the galleries and neighbouring forests. These predators have been hunted mercilessly and are now endangered species.

The caatinga

In the semiarid northeastern region the scrub is known as the caatinga, from an Indian term meaning "white forest." Covered in thorns, these woods may in places attain heights of up to 30 feet with interlocking branches that present the *vaqueiro* (cowboy), who is usually clad in protective leather garments, with an almost impenetrable barrier. Where water is scarce, the caatinga may be low, widely spaced, and intermingled with cacti. Somewhat moister areas known as the *agreste* support light woods; these are located principally between the caatinga and the coast and are at higher elevations where most of the remaining moisture is wrung from the trade winds.

Amazonia. In contrast to the scrublands that border it to the south, the Amazonian rain forest and the waters that drain it have a remarkable abundance of plant and animal life, even though the ecosystem is a fragile one that can easily be destroyed. While giving the impression of monotony because of the apparently similar tree crowns that rise to a more or less uniform height of around 150 to 200 feet, these lands contain the greatest variety of plant species on Earth—many thousands, less than half of which have been identified. Furthermore, serving as a defense mechanism against blight and other natural enemies, individual plants of any species tend to be widely dispersed. A typical acre of forest in the Amazon may contain 250 or more tree species, compared, for example, with perhaps a dozen in any acre of woods in the northeastern United States.

The rain forest

Because the crowns of these forest giants form a virtually closed canopy, allowing 10 percent or less of the sunlight to filter through, there is little plant or animal life on the ground below. The trees, however, are festooned with a wide variety of epiphytes, bromeliads, and lianas, while their branches teem with animal life, from insects, snakes, and tree frogs to numerous species of monkeys and a bewildering variety of birds. Some 1,400 bird species have been identified in the immediate vicinity of the main stream of the Amazon alone, both in the treetops and within the complex ecosystem of the *várzea,* the riverine lands. There, along the riverbanks, are also found alligators, anacondas, boa constrictors, capybaras, and a number of lesser reptiles and mammals. In the waters themselves there are manatees, freshwater dolphins, and some 1,500 identified species of fish, with perhaps another 1,000 unidentified species; the variety includes many types of piranhas, not all of them flesh-eating, electric eels, and some 450 species of catfish.

The Amazon is also home to the world's largest freshwater turtle, of the genus *Podocnemis,* extinct everywhere except there and on the island of Madagascar. Weighing an average of 150 pounds (70 kilograms), these turtles were once a mainstay of the Indian diet, and they remain a highly prized food source. Although the turtles were placed on the endangered species list, illicit turtle meat has continued to be an available delicacy.

Settlement patterns. The settlement of what is now Brazil began many thousands of years ago with the arrival of the first tribes of Paleo-American Indians, migrants from North America who were probably of Asian origin. Nomadic hunters and gatherers, they inhabited the less hospitable parts of the country away from the larger rivers. By the time of the European arrival a second group had evolved, known collectively as the tropical forest Indians. Outnumbering the nomadic Indians, they were skilled farmers and fishermen who occupied the best lands of the Amazon and Paraguay river systems and most of the coastal plains, making up the bulk of the more than 2,000,000 native inhabitants of Brazil at the time of the European arrival.

Coastal settlements. The first European occupants of Brazil, around the beginning of the 16th century, settled among the coastal Indian villages or at the trading posts that were established at Salvador and Cabo Frio. They exchanged hardware and trinkets with the Indians for brazilwood, which was used for making a rich, fire-coloured dye (*brasa* is the Portuguese word for "live coals"). By the second half of the 16th century sugarcane became dominant in the colonial economy, giving rise to a scattering of urban centres, among which Olinda and Salvador were the most important. By this time the coastal Indians had been decimated, and slaves from Africa were imported to work on the rapidly expanding plantations.

The Southeast: mining and coffee. The Southeast is Brazil's most densely populated region. During the first two centuries of Brazilian colonization little attention was paid to the nearly inaccessible and seemingly unproductive highlands, although parties of explorers, known as *bandeirantes,* did traverse them from time to time, capturing Indians for slaves and searching for precious metals and stones. Taking with them a few head of cattle, which eventually expanded into the herds that came to dominate the economy from the caatinga to the Pantanal, some of the *bandeirantes* settled down in the interior and continued their search for gold and diamonds. The first gold strike came in what is now Minas Gerais in 1695, where diamonds were also found in 1729, attracting many plantation owners from the northeast who brought their slaves. They spent money lavishly on building fine towns, such as Ouro Prêto and Diamantina, and also invested in small industries to supply the mines as well as the farms that, before long, were producing a surplus for export. With the development of roads over the Serra do Mar to the coast, and the transfer of the colonial capital from Salvador to Rio de Janeiro in 1763, Brazil's economic and political centre shifted from the northeastern to the southeastern part of the country. Subsequently, during the 19th century, this was consolidated by the wealth generated by the great coffee plantations developed in the Paraíba do Sul valley.

Developing the plateau. Rio de Janeiro's population had passed 500,000 by the time the slaves were fully emancipated in 1888, while the city of São Paulo, entrepôt for all of Brazil south and west of Minas Gerais, was but a modest town of 65,000. That situation changed as a flood of European immigrants flocked to Brazil; some worked as tenants on the coffee plantations that were expanding across the *terra roxa* soils of São Paulo and northern Paraná, while others established themselves on small freeholds along the southern coast and in the forests. Whereas the latter group remained physically and culturally isolated until after World War II, the newcomers in São Paulo played a key role in building the railroads and establishing the industries that were later to give both the city and the state their preeminence in the Brazilian economy.

Settling the outback. While the more southerly parts of Brazil were expanding economically, the northeast was stagnating and beginning to feel the effects of overpopulation and an archaic landholding system under which all the best coastal lands were in the hands of a few powerful landowners. This had driven the earlier population of mixed Indian and European ancestry, the mestizos, ever deeper into what was then the sertão, or outback. They first settled the *agreste* and then the caatinga, wherever water or a small patch of moist ground was available. Severe droughts in the 1870s and 1880s forced many of these people to abandon their land, and they were recruited to tap rubber trees in the Amazon to supply the growing demand for this product. Half a century was to elapse before there were new opportunities for massive migrations from the drought-prone, impoverished northeastern region. One of these opportunities was to provide the large amount of labour that was required during the post-World War

Migration to the southeast

II period to build the constantly expanding urban centres of the southeastern region, culminating in the construction of Brasília. At the same time, northeasterners began occupying the sparsely populated forest lands along the remote northern perimeter of the Brazilian Highlands, including Rondônia and Acre. There they have been joined by emigrants from southern Brazil, who were displaced by agricultural mechanization.

The Amazon region in the north remains the most underpopulated part of Brazil, with an environment resistive to commercialization and settlement despite government development plans intended to lure more migrants. There has been a damaging trend to the ecology, nevertheless. Forestry and cattle raising, for instance, pose threats to the future of the rain forest and the Amazonian habitat.

The entire Amazon region had a population of only about 40,000 in the middle of the 19th century, but by the end of World War I this had risen, largely owing to an earlier rubber boom and the influx of northeasterners, to some 1,400,000; Belém and Manaus had grown from somnolent villages into modest-sized cities. In the late 1950s there was a minor boom along the Lower Amazon where Japanese settlers began raising jute and black pepper. The manganese deposits in Amapá were also being rapidly developed, and a new pioneer zone appeared along a highway between Belém and Brasília.

Urbanization. With its rural settlement patterns essentially defined, Brazil in the post-World War II period began a headlong drive toward industrialization, which turned the nation from an essentially rural society into one in which three-fourths of the population is urban. In 1940 less than one-third of the total population of 42,000,000 lived in urban areas; by the late 20th century more than that number lived in the São Paulo metropolitan area alone. Rio de Janeiro has the second largest metropolitan population, and other major metropolitan areas include Belo Horizonte, Salvador, Fortaleza, and Brasília. Somewhat smaller are Recife, Curitiba, Pôrto Alegre, and Belém. This rapid growth has led to a series of physical and social problems, while the demand for housing has raised land values to staggering heights. As a result, members of the middle class have been forced more and more to live in minuscule apartments in densely packed high-rises, while the poor are confined to shantytowns, so-called *favelas,* or in more distant developments that may be several hours away from the workplace. (P.E.J./R.P.Mo.)

THE PEOPLE

Ethnic groups. In Brazil the peoples of mixed race are increasing, while those of separate racial stocks are declining. The country has long been a melting pot for people of all races and a wide range of cultures. Although social interactions have not always been without strife and exploitation, the tolerance of the Portuguese Brazilians for other peoples and official policies that favoured assimilation have minimized conflicts. The Portuguese language and the Roman Catholic religion were other unifying factors. Where assimilation has not taken place, it has generally been regarded in Brazil as the choice of a particular group and not the result of any exclusionist government policy or reaction by the population. A case in point is the intransigence of a small minority of the some 200,000 remaining Indian tribal peoples, who have persistently rebuffed overtures of the National Indian Foundation.

Indians. Although none of Brazil's native Indian groups achieved the high levels of civilization and wealth of those in the Andes and Central America, the tropical forest peoples were, nevertheless, superbly adapted to their environment. From them, the Portuguese traders and early settlers learned techniques of tropical crop cultivation, especially the growing of corn (maize) and cassava, a starchy root that remains a Brazilian staple. They learned to sleep in hammocks, which are still preferred to beds by many in the more northerly regions, and to make dugout canoes and the sailing rafts, jangadas, that continue to be used along the northeastern coast. Perhaps one-third of all Brazilians have some Indian ancestors, a fact that is particularly apparent among the people living within and to the west and north of the northeastern region.

As long as the Portuguese settlers limited their activities to barter, their relationship with the Indians remained peaceful. But it changed quickly when the colonizers decided to use the indigenous population as manual workers. Incapable of submitting to the economic and social constraints of the plantation system, the Indians died or fled. They took refuge in the most distant and inaccessible areas—in the forest regions between the rivers of the Amazon Basin or in the savannas of the Mato Grosso. They did not, however, find themselves completely sheltered there, for, from the 16th to the 18th century, the Portuguese organized *bandeiras* (expeditions or raids), launched from São Paulo or the northeastern towns, to hunt down Indians and enslave them. These raids contributed to the decimation of the Indian tribes, as did simple contact with whites, since the Indians had no immunity to European diseases (influenza, measles, smallpox).

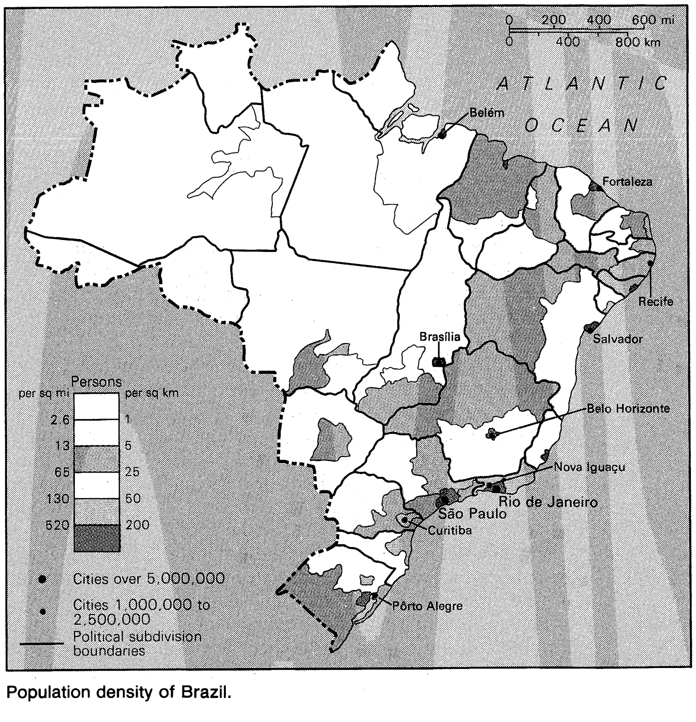
The flight to the interior

Blacks. The slave trade, between the 16th and the 19th centuries, brought 3,000,000 to 4,000,000 Africans to Brazil. The principal groups came from West Africa and from Angola. Unlike the Indians, the Africans adapted as labourers and domestics. Some, however, fled into the interior. During the 16th and 17th centuries the majority of blacks went to northeastern Brazil and were employed in the sugarcane plantations. From the 18th century onward, when the mining of gold and diamonds began, they went to Minas Gerais. After the abolition of slavery in 1888, many blacks left the regions where they had worked as slaves, either for other agricultural regions or for the towns. The black population then became more dispersed, although northeastern Brazil still has the heaviest concentration.

Historic distribution of the black population

Africans and their descendants have had a profound effect upon Brazil's racial and cultural evolution. Miscegenation was more acceptable in Brazil than in most other slave-owning countries so that in many areas, especially along the east coast, people of mixed ancestry outnumber those of pure racial stock. African music, dance, food, and religious practices have become an integral part of Brazilian culture. Nevertheless, the darker-skinned members of the population, although there are many exceptions, tend to be at the lower end of the economic hierarchy, with the whites at the top. This situation is said to be due to the perpetuation of generations of blacks in regions that have fewer economic and social opportunities. Their difficulties have been further compounded by competition from the European and Asian immigrants who have arrived in Brazil since the 19th century seeking greater opportunities.

Europeans and other immigrants. European elements



Population density of Brazil.

constitute the majority of the Brazilian population, a result of the steady influx of the Portuguese to all regions. For three centuries Lusitanian (*i.e.,* Portuguese) immigrants were practically the only Europeans to enter Brazil, and no limitation has ever been placed on their coming. They were found in all classes of society and were anxious to obtain wealth quickly in the plantations or in commerce.

It was only after the proclamation of independence in 1822 that immigrants of diverse origins were added to the Portuguese. The most numerous of the non-Portuguese European immigrants were Italians, who settled in large numbers particularly in São Paulo and northern Rio Grande do Sul. Similar to the Portuguese in many respects, the Italians were easily assimilated. The less numerous immigrants from other Mediterranean countries, including those from Spain and from Middle Eastern countries such as Syria and Lebanon, also adapted rapidly to their new homeland. With the Italians, they have made important contributions in industry, finance, politics, and the arts.

German and Japanese colonies

Immigration of Germans in the 19th and early 20th centuries and of Japanese after World War I added further diversity to the ethnic mix. In contrast to the Mediterranean peoples, however, it took the Germans and Japanese three to four generations to become assimilated. Besides the wider cultural gap that separated them from the general population, they were for many years physically isolated in rural areas. The cultural isolation of these groups was encouraged by their home governments, which had sponsored colonies and continued to supply them with teachers, and textbooks in their native languages. Since World War II, however, the people of German and Japanese descent have been largely integrated into the mainstream of Brazilian society. Other ethnic groups include Slavic peoples from eastern Europe. By the late 20th century there was little immigration, and less than 1 percent of Brazil's population was foreign born.

Language. Portuguese has undergone many transformations, both in the mother country and in its former colony, since the language was first introduced into Brazil in the 16th century. Although the two countries have, from time to time, standardized their spelling so that the written word remains mutually intelligible, pronunciations, vocabularies, and the meanings of words have diverged so widely that, it has been said, it is easier for some Brazilians to understand films in Spanish from other Latin-American countries than those from Portugal. New words and expressions in Brazilian Portuguese have been introduced by Italians, Germans, Japanese, and other immigrants and from across the borders with Spanish-speaking countries. One notable example is the universal use in Brazil of *tchau,* for farewell, adopted from the Italian *ciao.* Other words have entered through contact with foreign products and technologies.

Some authorities, however, suggest that the greatest divergence of the Brazilian language from the Portuguese goes back to contact with the Indians. The principal language spoken by the tropical forest peoples of Brazil, Tupian, or Tupí-Guaraní, became the lingua franca between the natives and the Portuguese traders, missionaries, adventurers, and administrators; it continued to be used similarly in the Amazon and western Brazil until the 19th century. The Tupian influence in Brazilian place-names is overwhelming, and it is estimated that thousands of indigenous words and expressions have entered the Brazilian Portuguese language. More generally, as a result of the Tupian influence Brazilian Portuguese became more nasal than that of the homeland, and Brazilians generally speak more slowly, pronouncing all the vowels.

Religion. Almost 90 percent of the Brazilian people are Roman Catholics, making it the largest country of that faith in the world. Roman Catholicism, however, ceased to be the official religion after the proclamation of the republic in 1889, causing the close links that existed between church and state to loosen. But the predominance of Latins among the immigrants of the 19th and 20th centuries has contributed to the lasting strength of the religion. Among the other religions, the majority are Protestants, dominated by fundamentalist and Pentacostal sects. There are also small numbers of adherents to Eastern Orthodoxy, Buddhism, Shintō, Islām, and others, all of which together are about numerically equal to the Spiritualists, whose beliefs overlap those of other religions, including the animism of the Indians.

African cults

A large number of Brazilians temper their Christianity with beliefs imported from Africa, as represented by such cults as Macumba, Candomblé, Xangô, and Umbanda. Among the more influential of these was Candomblé, which became the predominant cult of Bahia. The Nagô Candomblé sect, derived from the religion of Yoruban slaves, became particularly strong, and some of its rites were adopted by other sects and cults. Macumba and Umbanda are most widely spread in Rio de Janeiro state and Xangô in Pernambuco. The cults in general identify their deities with the saints of Roman Catholicism, and they believe that the deities intercede for them with a supreme being. Animal sacrifices are a common element of cult rites. Priests and priestesses are mostly blacks, but adherents, especially in urban centres, are drawn from every racial group and social class.

Demographic trends. For demographic and other statistical purposes the government has divided the country into five major regions: the North, which includes the states of Acre, Rondônia, Amazonas, and Pará, and the territories of Roraima and Amapá; the Northeast, containing the states of Maranhão, Piauí, Ceará, Rio Grande do Norte, Paraíba, Pernambuco, Alagoas, Sergipe, and Bahia, and the island territory of Fernando de Noronha, some 225 miles off the Atlantic coast; the Southeast, made up of the states of Minas Gerais, Espírito Santo, Rio de Janeiro, and São Paulo; the South, including the states of Paraná, Santa Catarina, and Rio Grande do Sul; and the Central-West, consisting of the states of Goiás, Mato Grosso, Mato Grosso do Sul, and Tocantins, and the Federal District, in which Brasília is located.

Population characteristics. Like most developing nations, Brazil's population is young; the percentage of people under 20, however, declined to less than half of the total population between the 1970s and the 1980s, while older age groups increased. The modernization of society has equally been a factor in the increase in life expectancy and a lowering of the birth rate, along with a general lowering of the population growth rate. Despite such changes the infant mortality rate has remained high. This rate, however, varies widely from the affluent urban districts, where it is quite low, to the *favelas* and other poor communities, particularly in the Northeast, where it is much higher.

Birth rates

The number of births to women of childbearing age (15 or over) has declined but continues to show significant regional differences. In 1940 the national average was just over six births per female of childbearing age, with a high of eight to 8.5 in the most rural states. The lowest figure in 1940 was in Rio de Janeiro state at 4.5, but by the 1980s the national average had dropped even below that figure. This decline has been attributed to improved health care, the gradual acceptance of family planning, including birth control measures, and, possibly, the declining influence of the Roman Catholic Church.

Migration. Migration from one part of the country to another has always been a feature of Brazilian society, especially when new frontiers or opportunities have presented themselves. During the late 20th century the western part of the country had the highest net influx of population, attracted especially to the Federal Capital and Rondônia. Indeed, the main purpose in creating Brasília in the country's centre was to instigate development of the hinterland. The Southeast has also received large numbers of migrants, but these have been very unevenly distributed: São Paulo and Rio de Janeiro have made notable gains, whereas Minas Gerais and Espírito Santo, the most rural southeastern states, have had steady reductions of their population through migration. The Northeast has suffered significant population loss to emigration, but this has been evenly distributed across the region except for being higher in Piauí, in the heart of the drought region, and in Maranhão, where it was much lower. The latter is somewhat of a geographic anomaly because its western half lies within the Amazonian rain forest while its eastern half is in the drought quadrilateral; thus migration into the former just about balances the exodus from the latter.

Population growth in the South

Although the North and the Central-West regions have the strongest percentage of population growth in Brazil, the South, and São Paulo in particular, has benefited significantly from substantial increases, both from internal migration and from foreign immigration. The South is the richest, the best equipped, the most industrialized, and the most vital region of Brazil; it receives the greater part of the flow of international immigration. Foreign technicians and skilled workers have generally achieved a comfortable standard of living there.

Domestic migrations are, in effect, motivated primarily by the lowering of rural income and the persistence of an archaic agrarian structure, as well as highly variable climatic conditions. Whole families of poor *sertanejos,* and also farmers from the *mata* and the *agreste,* flee toward the towns, the population of which is thus swelled by a mass of unqualified, underemployed people with a very low standard of living. A strong movement, therefore, carries Brazilians toward the cities and accelerates the mixing process in a nation becoming increasingly urban. It also induces—notably among the poorest people—a consciousness of regional economic imbalances and of excessive social inequalities. (M.Ml./R.P.Mo.)

THE ECONOMY

The sheer extent of Brazil's primary resources has made its economy, despite its relative lack of development, one of broad international significance. It is one of the world's leading agricultural nations and is especially well-known as the world's most prominent coffee-producer. Brazil is also important for what it has not yet fully exploited—its vast mineral and hydroelectric potential, its hardwood forests, and millions of acres of soil, most of which could be fertile given adequate water and fertilizer. As its manufacturing sector develops, Brazil also has taken a significant place among the world's industrial producers, its iron ore production having grown to a high world ranking. The city of São Paulo has become one of the world's major industrial and commercial centres.

Boom-and-bust economy

Despite the vastness of its land and the richness of its resources, Brazil's economic history can be largely characterized as a string of boom-and-bust periods. This was the result of being oftentimes heavily dependent on one or two major agricultural products, the markets for which were highly sensitive to fluctuations in the world economy and politics. This cyclical aspect of the economy began with the export of brazilwood in the early colonial times and was continued with the sugar boom of the 16th century, the mineral boom of the 18th century (paced especially by gold and diamond mining), the coffee boom beginning in the mid-19th century, and the rubber boom of the late 19th and early 20th centuries. In reaction to this background of economic fluctuation, based on a largely agricultural economy, Brazilian governments of the 20th century have placed heavy emphasis on programs designed to diversify the nation's production. In the process, industry has been heavily encouraged, and by the late 20th century manufacturing accounted for the largest single sector of the gross domestic product. The economy has gone through periods of remarkable growth since the mid-20th century, but serious economic problems persist and have been aggravated by political turmoil. Income distribution continues to be uneven, as it was in colonial times, with the large poor segment receiving the smallest share. Inflation, at times skyrocketing, coupled with periods of recession and one of the world's largest foreign debts became characteristic elements of the nation's economic difficulties in the late 20th century.

Resources. Brazil has an abundance of resources, which, with suitable management, could continue to yield generously for generations to come. Improved transportation has made more of these resources accessible either for export or for use by Brazil's burgeoning industries and growing population.

Mineral resources. Brazil is known to contain extremely rich mineral deposits. It has vast iron ore reserves located mostly in Minas Gerais and Pará, the latter being also the site of extensive bauxite reserves. Of the known manganese-ore deposits, most are in Mato Grosso and Amapá. Low-quality coal resources are located in Rio Grande do Sul and Santa Catarina, but only a relatively small percentage—found in the latter state—is suitable for processing to produce metallurgical coal for steel manufacture. In addition, the country has reserves of a wide variety of other metallic and nonmetallic minerals, some of which provide important exports. Brazil has oil and natural-gas reserves, which became substantial with the development of offshore fields. The mineral reserves are under constant exploration by government and private sources.

Oil and natural gas

Biologic resources. About two-thirds of the country is under forest, providing one-seventh of the world's total forest area. Hardwoods predominate in the Amazon and Atlantic coastal zone. Exploitation of the Amazon rain forest, however, has been hampered by the inadequacy of transport facilities. Arable land is not abundant, considering the size of the country. With a coastline of more than 4,600 miles and numerous well-stocked rivers, Brazil has access to substantial fishing grounds. This resource, however, is underdeveloped and productivity is low.

Hydroelectric resources. With one of the most extensive river systems in the world and with plentiful rainfall in all of the country except the Northeast, Brazil has one of the largest hydroelectric potentials in the world. Installed generating capacity is concentrated in the Southeast and the South, which consume most of the power produced in Brazil.

Since 1950 there has been a rapid expansion in total power capacity, mainly through hydroelectric development. The government has given high priority to harnessing the country's substantial hydroelectric resources, as the scope for expansion in thermal capacity is limited by the poor quality of Brazilian coal. Brazil's first nuclear reactor, Angra I, began operation in 1982 near Rio de Janeiro, and the Paraná River's Itaipú hydroelectric complex, the world's largest power station at its completion, began operating in 1984.

Agriculture. Agriculture has been declining since the mid-20th century in its share of the gross domestic product and in the number of its workers; less than one-third of Brazil's labour force is employed in farming and stock raising. The country is essentially self-sufficient in basic foodstuffs and is a leading exporter of a wide range of tropical crops. Unlike most of the other Latin-American countries, Brazil has increased agricultural production by greatly enlarging its cultivated area since World War II. Yet only about 9 percent of the land area is cultivated, much of it in small holdings using relatively primitive technology; another 20 percent is used for stock raising.

Mechanized farming is still somewhat rare in Brazil, and it is mostly employed in the South and Southeast. Few tractors are available in the Northeast although that region contains about half of Brazil's farms; most of these are less than 12.4 acres (five hectares) in size, however. Some large-scale government irrigation projects have been built in the Northeast, but they were costly and support only a limited number of farm families on individual plots or on cotton plantations.

Coffee growing

Brazil is the world's leading producer of coffee, which is its most important single export. São Paulo and Minas Gerais are the principal coffee-producing states, followed by Paraná and Espírito Santo. Soybeans and their derivative products, particularly animal feeds, have become another important export. Most of the country's soybeans are grown in Paraná and Rio Grande do Sul, but mechanization and a high crop value have made it economically feasible to fertilize the savanna soils, expanding production beyond the *terra roxa* region and making Mato Grosso do Sul a leading soybean state.

Older segments of Brazil's agriculture have also responded to economic opportunities. An example of this was the rapid expansion of sugarcane acreage, mostly in São Paulo but also along the northeastern coast, under the government's program to replace gasoline, made from costly imported fuel, with ethanol (ethyl alcohol). Processing the sugarcane in modern plants became the most successful such program in the world, and for a number

Paraná pine growing in a field of soybeans in the *terra roxa* region of Rio Grande do Sul state, southern Brazil.

Karl Kummels—Shostal Assoc.

of years virtually all new automobiles in Brazil have been engineered to run on this fuel.

Some ethanol is made from cassava, of which Brazil is the world's largest producer. It is also usually a world leader in growing edible beans, castor beans, cacao, and bananas and is the largest grower of rice in the West. Although the bulk of these products are consumed domestically, varying amounts are exported, as are such crops as jute and black pepper from the Amazon region; palm oils from the northeast coast; garlic from Minas Gerais; peanuts (groundnuts), oranges, and tea from São Paulo; and tobacco from Santa Catarina and Rio Grande do Sul. The latter state is also the centre of Brazil's meat-packing industry; vast grazing lands there and elsewhere in Brazil have given rise to one of the world's largest livestock populations.

Forestry and fishing. Brazil's abundance of plant and animal life is particularly identified with the Amazon region, yet only a small portion of the annual timber harvest comes from northern Brazil, and much of that is used for fuel. Most of the small timber yield of the Northeast is also used as fuelwood. The forests of western Maranhão and southern Bahia are important sources of charcoal, as are those of Tocantins and eastern Minas Gerais, the latter accounting for the largest share of Brazil's production of that fuel. The largest part of Brazil's timber production comes from the South and Southeast, about half of it from plantations of eucalyptus, originally introduced from Australia, with additional yields from Honduras pine and several other exotic species. Most of the harvest from these plantations is used in the manufacture of cellulose and paper products.

Fishing

Brazil has a significant freshwater fish catch, of which more than one-fourth comes from the Amazon and its tributaries. Another large segment of the country's freshwater catch is from the Northeast, much of it from reservoirs that the government has stocked with tilapia, a fast-growing fish introduced from Africa.

Whereas most of the freshwater fishing is done by individuals, more than two-thirds of the ocean catch is brought in by company-owned commercial fishing vessels. Most of these sail from Southern and Southeastern ports. In the Northeast company boats account for a considerably smaller percentage of the saltwater catch, most of which comprises lobster and shrimp caught primarily for export. In Fortaleza manufacturers use *cambulu,* a saltwater fish, and tilapia skins to make shoes, clothing, and accessories, products formerly made from the hides of alligators, which are now endangered species.

Industry. *Mining.* Almost every Brazilian political unit receives some revenue from the extraction of its minerals. The leading mining state is Minas Gerais, which, excluding oil and gas production, receives the largest single share of the country's mining revenue. This state has long been Brazil's major producer of iron ore, for internal consumption and export, from high-grade hematite found in Precambrian formations that extend from near Belo Horizonte eastward to Itabira's "Iron Mountain" in the upper Rio Doce valley. Minas Gerais' iron ore exports have been challenged by those from the huge Carajás Grande mine, located west of the Araguaia River in Pará at the end of a specially constructed, 500-mile-long railway to a port near São Luís. Most of Brazil's iron ore exports are in the hands of Brazil's giant, government-controlled mining and shipping company, the Rio Doce Valley Company. The state has controlled a number of other mining companies, but a move to denationalize them began in the mid-1980s.

Except for aluminum and manganese from the Amazon region and iron ore, Brazil's industries absorb most of its other mineral production. This includes chrome, magnesium, and quartz from Bahia; copper and lead from Bahia and Rio Grande do Sul; asbestos from Goiás; and nickel from Goiás and Minas Gerais. The latter state also supplies Brazil with zinc and graphite, in both of which the country is largely self-sufficient. Brazil is self-sufficient in cassiterite (tin ore), found along a belt south of the Amazon River, while nearly all of the country's tungsten requirements are met from Rio Grande do Norte and those for silver by Bahia and Paraná. Coal production, which is centred in Santa Catarina, supplies more than half of the country's needs. Brazil is a major gold and diamond producer, supplied mainly by Minas Gerais, but quantities fluctuate widely from year to year and place to place as deposits are located and exhausted. Minas Gerais, Bahia, and Espírito Santo are the major sources of Brazil's enormous range of gems—topazes, amethysts, opals, aquamarines, tourmalines, emeralds, and others—that make Brazil the largest world supplier of semiprecious stones.

Gemstones

Petroleum. Virtually all of Brazil's oil and natural gas fields derive from deposits laid down at various stages along a rift that developed as the South American continent began its separation from Africa. Of these, the offshore marine shales are the most productive. For the first 25 years after commercial oil production began in Brazil, in 1940, it was confined to the Bahia Basin just north of Salvador. This expanded subsequently to include offshore and onshore fields that extend from Fortaleza to Santos, although virtually all of the country's natural gas comes from Bahia and Sergipe states. In 1987 the first commercial oil well in the Brazilian Amazon, some 450 miles southwest of Manaus, began production.

More than half of Brazil's petroleum production comes from the Campos Basin on the continental shelf off Rio de Janeiro state. There Petrobrás, the government-controlled petroleum company that was established in 1953, has developed some of the most advanced deep-water drilling technology in the world. This has included test drilling on the continental slope into formations under 1,200 to 3,000 feet of water. The higher-yielding finds at this site have added considerably to the country's proven reserves.

Manufacturing. Since 1967 Brazilian industry has contributed a greater amount to the gross national product than has agriculture. Manufacturing is the most dynamic sector of the economy. Although it developed principally to serve the domestic market, Brazilian industry exports a wide range of manufactured goods.

Much industrial development that followed the Great Depression of the 1930s was the result of nationalism and developmental policies spearheaded by state corporations. Investment capital was then in short supply unless it was accompanied by foreign control, which Brazil wanted to avoid. As a result, the government took control of some of the country's largest industrial enterprises, usually in partnership with one or more local or foreign corporations. It gained the right to sell stock in government companies to private investors, which was the case with Petrobrás and the Rio Doce Valley Company.

Although government involvement in the industrial sector has been criticized for promoting political and social objectives rather than economic ones, and for its cumbersome bureaucracy and inefficiency, some thriving industries can attribute their successes in domestic and international markets to government measures. Besides direct investments, these have included tax and other incentives, protective tariffs, import restrictions, and the outright prohibition of imports.

An example of such success is the motor vehicle industry, which was established in the 1950s to replace U.S. and German imports and assembly plants. Begun with a joint government–Alfa Romeo truck factory and a series of incentives and import tariffs on automobiles, Brazil was producing more than 1,000,000 vehicles a year by the late 1980s, some of which were exported.

Another such industry is petrochemicals; the largest in Latin America and a major factor in the value of the country's industrial production, it was started by Petrobrás and subsequently opened up to national and foreign investors. The government was also instrumental in starting a modern shipbuilding program that, with private capital as well, produces various types of vessels, some of which are sold abroad. Embraer, begun with government support, built small airplanes in the days of poor road conditions and now manufactures and exports a wide range of aircraft. Government protection has also been instrumental in establishing a burgeoning microelectronics and personal computer industry.

One notable exception to Brazil's manufacture of goods primarily for domestic consumption is its armaments industry, which exports most of its production to countries throughout the world and places Brazil among the world leaders in arms sales. Arms-producing companies may be under private or government control. Their products are reputed to be of good quality, moderate complexity, and reasonable cost, making them especially suitable to their main market, Third World countries.

Textiles and clothing

Not all of Brazil's successful industries involve government participation. An example is the textile industry, which began in Bahia in 1814 and was based on local supplies of raw cotton and is now centred in São Paulo. Of equal significance is the manufacture of clothing and footwear; of the latter, Rio Grande do Sul produces most of the exports. This industry began in the 1820s with small leather works based on a surplus of hides from the meat-packing industry but now imports a portion of its raw materials from Argentina and Uruguay. Brazil's metallurgical and engineering industries originated in Minas Gerais and Rio de Janeiro during the 18th-century mining boom and have become one of the chief employment sectors in manufacturing.

With few exceptions, the Southeast contains the largest, the most varied, and the most efficient establishments in every sector of industry. It also employs the majority of the country's industrial workers, who earn most of Brazil's wages and produce the largest value of its goods; São Paulo state alone is responsible for about half of Brazil's industrial production. The next largest percentage of the nation's industrial workers, about one-fifth of the total, are in the South. The Northeast, by contrast, employs only about 10 percent of Brazil's industrial workers, and their wage scale is lower than those in the Southeast and South. Generally speaking, Brazil's factories are not large; only a very small percent employ a hundred or more workers. As might be expected, the largest firms are in the Southeast, with the next largest in the South.

Services. The service sector, including private and government services, is not only one of Brazil's largest employers, absorbing more than one-fourth of the work force, but is also its fastest growing one. The private sector has been expanding at a somewhat faster rate than the public sector. Government services include not only the usual array of national and local bureaucracies and public services but also a host of special purpose agencies. Among the largest of these are the superintendencies for the development of the Northeast, SUDENE (1959), and of the Amazon region, SUDAM (1966). These agencies funnel federal funds into development projects and oversee the tax incentives that are intended to stimulate investments of private capital, especially in manufacturing and rural enterprises. Although the overall objectives of the various government agencies remain fairly constant, specific applications may vary significantly from one administration to another. Furthermore, their functions frequently overlap, especially at the local level. In the private sector, the largest number of workers is employed in the hospitality industries (hotels, restaurants, and bars) and in repair shops of various kinds. Retail sales and personal services each account for most of the rest of the private sector workers.

Tourism. Tourism is a growing industry in Brazil, but the concept of building resorts for the exclusive use of tourists is an unfamiliar one. It has been limited for the most part to a few luxury hotels, mostly in Rio de Janeiro, that cater to foreign visitors, some long-standing spas and resorts mostly in the Minas Gerais highlands, and a facility at the relatively isolated Iguaçu Falls. The major attractions are located in or around urban centres with well-established hospitality industries and are, therefore, easily accessible. Also, the proportion of foreigners traveling to Brazil who might require special facilities is still not large compared with the number of nationals who travel within their own country, using the same type of facilities as they would for family or business visits. The larger Brazilian cities all have a wide range of accommodations. A growing number of Brazilians travel throughout the country by automobile, although during vacation periods airlines are also heavily booked.

Finance and trade. The national Monetary Council under the direction of the minister of economy is Brazil's chief financial policy-making body. It controls the Central Bank of Brazil, which issues currency and controls the money supply, credit, foreign capital, and other top-level financial matters. The federal government uses other financial institutions to implement its policies as well, the most important of which is the Bank of Brazil. It is not only the largest bank in the country, with numerous agencies at home and abroad, but is also the main source of long-term loans in such areas as rural credits and the financing of manufactured goods exports. The National Economic and Social Development Bank channels government and international loans into large-scale development projects in industry, mining, agriculture, and the infrastructure, including loans to state governments for projects that they cannot finance themselves, such as the São Paulo and Rio de Janeiro subways. The National Housing Bank provides home-building loans, and the Federal Economic Bank makes short-term loans to individuals.

Each state has its own government bank, among which the Bank of the State of São Paulo is the most important. A lesser share of Brazil's commercial banking is in the hands of private banks, which also provide short-term loans and the popular savings accounts, Carnets de Poupança, that guarantee interest yields above the rate of inflation; these are also offered by the Bank of Brazil and the state banks.

Inflation

Almost continuous high rates of inflation in the late 20th century have affected every aspect of Brazil's economic life, which is also subject to the reactions of a volatile, underregulated stock market. A number of similarly unstable commodity exchanges also operate in various cities. Outside of the market, virtually all transactions are indexed for inflation, based officially on the constantly corrected value of the government's bonds. Such corrections apply not only to savings accounts but also to tax payments, calculations of assets, debt payments, selling prices, and the like. Separate monetary corrections are used for regulating food, transport, rents, and salaries, which, in turn, determine pensions. Because the latter group may be changed arbitrarily to reflect fiscal or political policies rather than social needs, they can be the source of considerable popular unrest even though this tends to be unfocused and ineffective.

High inflation rates in Brazil came, at least in part, from heavy government financing of industrial expansion, the need for government financial institutions to subsidize business loans, and a great increase in foreign borrowing as internal credit became extremely tight. Various government plans to curb inflation, some of which introduced a new basic monetary unit, have met with limited success.

Foreign trade has been critical to the Brazilian economy throughout the nation's history and continues to play a significant role. Although exports account for only a small part of the national income, their importance is understandable in the context of the nation's need to import capital goods and raw materials, particularly crude oil. The historic dependence on agricultural products for exports has changed, however, as manufactured wares have grown to take on an increasingly greater share of the market since the mid-20th century. Brazil traditionally has had difficulty maintaining a favourable trade balance, but in the 1980s, after a period of deficits, a trend toward surpluses was established. Much of this was negated, however, by the huge foreign debt payments that had been incurred.

(E.I.U./R.P.Mo.)

Trading partners

The largest single share of the value of Brazil's imports is accounted for by petroleum from the Middle East, which is the destination for only a small share of the country's exports. The United States is the largest customer for Brazilian goods, and that country ranks with the Middle East as an exporter of products to Brazil. Other imports originate from the countries of the European Economic Community and from the other countries of South America as well as the Far East and Africa. In addition to the United States, Brazil's exports go to the European Economic Community, with the Far East and other countries of the Americas absorbing most of the remainder.

Transportation. For a country as large as Brazil, development of an efficient means of transportation has been a matter of critical importance. Throughout much of its history the nation's different regions remained isolated from one another, but this changed dramatically after World War II, first with the growth of air transport and, two decades later, with the extension of a modern road network. Most passenger and freight traffic moves over the highways, with coastal and inland shipping the next largest mode of transportation.

Roads. At the time of the construction of Brasília in the late 1950s, the number of good paved roads in Brazil was limited. A four-lane highway linked Rio de Janeiro and São Paulo, but there were no paved roads from those cities to Pôrto Alegre, Curitiba, the Northeast, or west of Belo Horizonte. During the rainy season surface communications could be interrupted along the major regional links for a week or more at a time, stranding motorists in areas with limited housing and food supplies.

The construction of Brasília, for which many bulky materials had to be airlifted in during the rainy season, alerted the country to the poor state of its roads, and when the military assumed power in 1964, the upgrading of the road system became a primary objective. As a result, an excellent system of paved roads now connects all of the major points in Brazil, including several links northward into the Amazon region.

Air transport. Brazilians were among the pioneers in flying, and they have long claimed that their countryman Alberto Santos-Dumont, not the American Wright brothers, flew the first successful airplane. Numerous airlines flourished in Brazil at one time or another, but they have been consolidated into three major ones that compete nationwide: VARIG, which since the late 1920s has been a largely employee-owned airline and is a major carrier within Brazil and also the nation's main international air service; VASP, the São Paulo State Airline, which is largely local but also has scheduled flights to the United States and the Netherlands Antilles; and Transbrasil, which is a privately-owned airline that provides mostly domestic service.

Major airports

Every capital and important city in Brazil has a major airport, and most of the smaller cities are served by jet aircraft. Few locations are without at least a dirt landing strip. Although São Paulo, Rio de Janeiro, Belo Horizonte, and Brasília are all linked by shuttle services (*ponte aérea*), overall flight frequencies and the size of terminals are everywhere at much smaller scales than at comparable centres in western Europe or North America. This is due to the relatively high cost of air fares and competition from inexpensive intercity bus services. Rio de Janeiro and São Paulo handle most international air traffic.

Railways. Except for the commuter lines to the Rio de Janeiro and São Paulo suburbs, and certain bulk ore carriers, railroads are only of minor importance in Brazil's transport network. In contrast to Brazil's dynamic highway construction program, few new railways of any significance have been built in the country since World War II, when Rio de Janeiro was linked by rail to Salvador because German submarines were taking a heavy toll of coastal shipping. The minimal construction since that time includes a branch line from Minas Gerais to Brasília and the ore-carrying line inaugurated in 1985 between the Carajás Grande mining project and Pôrto do Itaqui, near São Luis. Most of Brazil's track dates from the 19th century, when the economy was based on exports of raw materials. Because a number of different, incompatible gauges were used, integration into a national network was impeded.

Shipping. Brazilian coastal shipping was, for many years, in no better condition than its railways and, like them, was confined to carrying low-value bulk goods that could tolerate long delays and not infrequent losses. After the federal government launched a shipbuilding program in the 1960s, however, the tonnage of cargos increased markedly, and, more significantly, as the frequency and reliability of services improved, a larger percentage of higher valued goods was carried. Lloyd Brasileiro is Brazil's main shipping line, but Petrobrás and the Rio Doce Valley Company operate a significant number of special-purpose vessels, and there are several other smaller lines.

The upgrading of Brazil's shipping coincided with the construction of specialized bulk terminals, including one on Sepetiba Bay, west of Rio de Janeiro, and the Pôrto do Itaqui ore terminal. These were preceded by the iron ore terminal at Tubarão, near Vitória, and the oil terminal at São Sebastião, on the São Paulo coast. São Sebastião accounts for about half the value of the state of São Paulo's imports, a large portion of the rest being handled through the port of Santos, which has the nation's largest port activity. Other significant ocean ports include Rio de Janeiro, Paranaguá, and Recife.

River navigation

The extensive Brazilian river system has a total navigability of about 27,000 miles. Within the Amazon Basin navigable waterways are the principal means of transportation in northern Brazil, extending into the Spanish-speaking countries to the west. The two principal Amazonian ports, connected by a modern, if at times erratic, steamer service, are Belém, at the mouth of the river, and Manaus, some 1,000 miles inland. These and lesser ports are important as trade centres for the multitudinous variety of craft that ply the waters of the main rivers and some 1,000 navigable tributaries. Between its natural and man-made obstacles, the Paraguay–Paraná–Plata river system is little suited for navigation, although certain stretches were used for local transport in the early days of settlement. There is still some barge traffic, however, and the system provides access to the Atlantic Ocean through the Río de la Plata estuary in Argentina, but only for shallow-draft vessels.

The São Francisco River is navigable for about 170 miles from the sea, after which access is blocked to shipping by the Paulo Afonso Falls. Above these the river was, for many years, an important avenue of transportation; however, silting in its upper reaches due to deforestation and the diversion of its waters for a series of hydroelectric dams have diminished the river's usefulness for transportation. Of the shorter rivers flowing into the Atlantic Ocean, only the Paranaíba in the far north and the Jacuí in Rio Grande do Sul are navigable from the sea for any distance, and then only by shallow-draft or barge traffic.

(R.P.Mo.)

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Brazil is a federal republic divided into 26 states and one federal district. The constitution promulgated on Oct. 17, 1969, under Brazil's military leadership, continued in force after civilian government was restored in 1985, although some key amendments were made and a constituent assembly was formed. A new constitution was promulgated on Oct. 5, 1988, the eighth constitution since the country's independence in 1822. The product of

The constitution

a 559-member Constituent Assembly, it abolished many of the traces of the former military regime. As the fundamental law, the constitution defines the rights of the citizen and the functions of the executive, legislative, and judiciary powers. The president's power to legislate was abolished in the 1988 constitution, as were government censorship of art and the use of torture and extradition for political crimes. The voting age was lowered from age 18 to age 16, and a provision was made for a presidential election in November 1989.

According to the constitution, legislative power is exercised by the National Congress, composed of the Chamber of Deputies and the Federal Senate. Congress meets every year, in two sessions of four and a half months each. The constitution gives Congress the power to rule in all matters that come under the jurisdiction of the union, particularly those related to the political or administrative organization of the union and to fiscal policies. Congress also has the power to rule on international treaties concluded by the executive, to authorize the president to declare war, and to approve or disapprove federal intervention in the states. Bills approved by Congress can be totally or partially vetoed by the executive. Congress has 30 days to overrule a veto by a majority vote.

The Chamber of Deputies is composed of representatives of the states elected every four years by direct universal suffrage and secret ballot. The number of deputies is in proportion to the population of each state, but no state can be represented in the chamber by more than 70 or by fewer than eight deputies.

The Federal Senate is formed by representatives of the states (three from each state and the Federal District), who serve for a term of eight years. The Senate is renewed every four years, the first time by a third of its members and the second time by the other two-thirds. Senators are directly elected by the electorate of each state in a straight majority poll.

Executive power

Executive power is exercised by the president, who is assisted by ministers of state. From 1945 to 1964 the president was elected for a five-year term by direct universal suffrage and secret ballot. After the military took power in 1964 the president and the vice president were elected indirectly, first by the National Congress and later by an electoral college formed of members of Congress and representatives of state assemblies. A 1977 reform extended the mandate of the president to six years. In 1985, however, under a civilian government, an amendment was passed restoring direct presidential elections, establishing a four-year term of office for the president, and instituting other political reforms. The president's term was extended to five years in 1988.

The executive has wide powers, particularly in economic and foreign policy, finances, and internal security. The executive can decide, for example, that any of its bills submitted to the legislative branch shall be voted upon by Congress within 30 days. If Congress does not comply within this period, the bill is considered approved. The president is assisted by his cabinet, made up of the various ministers of state and several other heads of ministerial-level departments, all of whom are appointed by the executive.

Judiciary power

The judiciary power is made up of several institutions. The Federal Supreme Court is composed of 11 members named by the president with approval of the Federal Senate. It alone decides on litigations between foreign states and the different entities of the union, as well as on differences between the political or administrative divisions of the union. The court also deals with either the demands or writs of mandamus against acts of the highest authorities of the republic (the president, Congress, and judiciary) and decides on demands of habeas corpus already judged at a lower level. It also pronounces on the constitutionality of laws and decrees, and it judges common-law crimes by the president, ministers of state, members of Congress, or magistrates of higher courts.

The Federal Superior Court is composed of 33 judges appointed by the president with the approval of the Senate. It hears cases involving governors of the states and the federal district and members of the judiciary. It also decides on writs of mandamus and habeas corpus against ministers of state and members of its own court, decisions that were formerly the responsibility of the Court of Appeals. Under the 1988 constitution the Court of Appeals serves as the court of last resort for common pleas. Each state or territory, as well as the federal district, constitutes a judiciary section. The federal judges there instruct and judge in the first instance certain political crimes, crimes committed against the labour organization or crimes committed on the occasion of strikes, and cases related to public organizations. Electoral courts and judges are responsible for the registration of political parties and the control of their finances. They also fix the date of elections and prosecute and make judgments on any electoral crimes committed. Labour courts and their judges mediate in individual or collective conflicts between management and workers. Military courts and their judges mediate in cases involving the military.

Local government

The states and territories are subdivided into municipalities. The municipalities are autonomous politico-administrative units created by the states, on condition that they have a minimum of population and their own fiscal resources. The municipalities are governed by mayors (*prefeitos*) and municipal councillors, both of whom are directly elected. The federal district, Brasília, is the capital of the union. Its *prefeito* is named by the president with the approval of the Senate.

The states are ruled by autonomous laws and constitutions within the framework of the principles defined by the federal constitution. The union can intervene in the internal affairs of the states, in cases foreseen by the federal constitution. Each state has a governor and a legislative assembly. Governors and assemblymen are elected by direct vote with secret ballots.

The political party system in Brazil began to emerge in the 1940s under President Getúlio Dorneles Vargas, who established the Social Democratic Party and the Brazilian Labour Party to buffer his weakening administration. A number of other parties were organized and entered elections through the 1950s and early 1960s, but few of them developed any political power of note. Under the aegis of the military government that was established in 1964 all political parties that had been previously active were abolished in 1965. They were replaced by one government party, the National Renewal Alliance, and an opposition party, the Brazilian Democratic Movement. In 1979 the two official parties were abolished in favour of a new system that would allow participation of more parties, but still under quite restrictive regulations. After civilian government was restored in 1985 all political parties were once again legalized. (L.Ms./Ed.)

Education. The educational system in Brazil is overseen by the Ministry of Education and its policy-making body, the Federal Council of Education. Education is free and compulsory at the primary level (grades one through eight, starting at age seven) and is also free at the middle, or secondary, level (three grades and an optional fourth). A broad educational gap exists in Brazil between the higher income groups and the masses in the lower income groups. About 20 percent of the people are illiterate, and about two-thirds have only four years of education or more. An attempt to correct some of the inequities in the system was made through reforms instituted in 1971, but results have been limited.

Primary education

Brazil's pre-primary schools are about equally divided between state, municipal, and private institutions. Of the primary schools nearly one-half are located in the Northeast, although they enroll only about one-third of the nation's primary students; whereas the Southeast, with about 40 percent of the primary students, has one-fourth of the schools. The primary schools of the Northeast, and in the North and Central-West as well, are smaller, more dispersed, and in the hands of less qualified teachers than those in the South and Southeast. Furthermore, the schools in the former areas tend to be financed out of meagre municipal budgets, whereas in the latter they are predominantly state-supported. For example, in Bahia there are six times as many primary schools supported by municipalities than are funded by the state; in São Paulo the balance is reversed. Private schools generally account for a small percentage of primary school enrollments. The federal government also operates a small number of primary schools, mostly in the Amapá and Roraima territories and in the city of Rio de Janeiro, the latter a legacy from the days when it was the federal capital.

Only about 10 percent of the primary students in Brazil continue on to the secondary level, and, of these, most do not complete the three- to four-year program because many of the curricula involve short-term vocational training. Low overall secondary school enrollment is caused not only by the necessity of entering the work force at an early age (the census includes children as young as 10) but also by the fact that these schools are generally located in the larger towns. Attendance therefore can involve the considerable financial burden of living outside the home community. This is particularly true in the Northeast, where there are few secondary schools. About one-half of Brazil's secondary students are in the Southeast.

Higher education

Compared with developed countries, university enrollments are small in Brazil, whether in relation to the overall population or to that of the cities in which they are located; higher education remains largely the prerogative of the wealthy and of the more ambitious members of the middle class. Every political subdivision, be it state, territory, or the Federal District, with the exception of Fernando de Noronha, has at least one university, although in many cases these are limited to those established by the federal government. The largest of the national institutions is the Federal University of Rio de Janeiro, which has a modern campus located on an island in Rio's Guanabara Bay, near the international airport. The universities of Minas Gerais and Rio Grande do Sul are the next largest federal institutions, followed by those of several cities in the Northeast. The largest and most important state university, with the country's largest graduate student population, is the University of São Paulo. It has a main campus in the city of São Paulo and branch campuses around the state. Students may also attend a number of private institutions, many of them run by the Roman Catholic Church.

Welfare and health. Although Brazil is a homogeneous country relative to its overall cultural institutions, the social gap between its small privileged upper class and the masses at the bottom of the earnings scale is vast. Sandwiched in between them is a modest-sized middle class.

Because of inflation, salaries are expressed as multiples of the official minimum wage. Two-thirds of the working population earns two minimum salaries or less. About one-third of the northeastern work force earns only one-half the minimum, or less, compared with much lower numbers elsewhere. Food prices and wages do not rise synchronously, and the buying power of one-half of the minimum salary may mean as little as the equivalent of one loaf of bread and one quart of milk a day.

The majority of workers are covered by various benefits: health, unemployment insurance, retirement and severance pay, forced savings, and holiday pay. These are paid by the employer to the National Social Security Institute on the workers' behalf and can add significantly to labour costs. Brazil customarily spends a greater percentage of its gross national product on social services than it does on its military budget. There are, however, widespread complaints about the level and quality of benefits received. The majority of the working population leads a meagre life, a situation that is exacerbated in the cities by high housing and transportation costs.

Many of Brazil's health problems stem from widespread undernourishment and endemic diseases such as malaria, yellow fever, dengue, amoebic dysentery, tuberculosis, schistosomiasis, and the dread Chagas' disease, which is transmitted by the bite of an insect that infests the walls of wattle-and-daub houses. The majority of these, however, are largely absent at higher elevations and in the subtropical climate zones. Brazil's major research institute for tropical diseases, the Oswaldo Cruz Foundation, is located in Rio de Janeiro.

Although most endemic tropical diseases have been eradicated in the major cities, their reintroduction by migrants from infected areas has become a serious problem as far south as São Paulo. This has been exacerbated by poor sanitary and housing conditions, particularly among Brazil's millions of shantytown dwellers, or *afavelados,* who are concentrated in and around the larger cities. Programs at all administrative levels, supplemented by privately supported clinics, have been established for the purpose of improving health conditions in the *favelas,* particularly in prenatal and infant health care.

Hospitals

Most of the hospitals in Brazil are public institutions, only about one-fifth being private. The number of doctors per person increases progressively from the North and Northeast through the South and Central-West and reaches the highest ratio in the Southeast. In terms of services available, the greater proportion of the country's doctors and hospitals are concentrated in the urban areas. There is also a wide gap between the quality and promptness of services provided for those in the upper-income groups by the private sector and that available generally through public institutions, even though Brazil's physicians generally devote a part of their practice to public service. Numerous state and national agencies operate a variety of health care services, although often with limited programs. The one with the broadest mandate is the National Institute for Medical Assistance and Social Welfare, which operates hospitals, clinics, and laboratories in both rural and urban areas.

CULTURAL LIFE

The cultures of the Indians, the Africans, and the Portuguese have together formed the modern Brazilian way of life. By far the most dominant of these cultures is that of the Portuguese, from whom Brazilians acquired their language, their religion, and most of their traditional customs.

Indian and African influence

The influence of the aborigines and Africans is still apparent, however. Tupí-Guaraní, the language that was spoken by most of the Brazilian Indians, was adapted by 16th-century missionaries for teaching the catechism and for a time was spoken widely throughout the interior of northern Brazil by Indians, mestizos, Portuguese, and blacks. Many words in modern Brazilian Portuguese have their origin in the Indian language. The Indian contribution to Brazilian culture is, however, perhaps most apparent in the Amazon Basin.

African influence on the Brazilian way of life is strongest in the plantation region north along the coast from Rio de Janeiro. Particularly in Bahia, there are traditional dishes of African origin, such as *vatapá*—made of rice flour, coconut oil, fish and shrimps, red peppers, and many condiments. Evident in northern coastal cities are religious cults of African origin. African influence is also reflected in Brazilian popular music, especially in the rhythmic sambas.

Government involvement in virtually every aspect of Brazilian life has failed to lessen the individualism and initiative of the people. Despite numerous social and economic problems, Brazilians continue to be exuberant in their celebrations, their art forms, and in support of their favourite football (soccer) teams. Although among the oldest and most stable of the European-derived cultures in the Americas, Brazil is a country that adapts readily to rapid changes and new opportunities.

Daily life. The rapid, large-scale urbanization of Brazil during the post-World War II period has altered the lifestyle of the majority of the country's inhabitants. In most ways, large Brazilian cities differ little from others in the Western world, but their greater degree of crowding and large volume of pedestrian traffic may in some cases compare more closely with the cities of China than of North America.

Brazilians' family ties, both immediate and extended, generally remain stronger than in North America; family members customarily live in relatively close proximity to one another, holding frequent reunions or going to the family farm or ranch on weekends and holidays. However, this traditional system of kinship ties depends upon a certain degree of wealth and stability for its preservation, and it is no longer as strong as it once was, given the increased mobility and urbanization of the Brazilian people. In the *favelas,* various members of an extended family may occupy the same dwelling, but this may be of necessity in some cases, rather than tradition.

The arts. Brazil has had many world-renowned literary figures whose cumulative writings are regarded by many to be richer than Portugal's because of their variety of racial and regional themes. Best known in the 19th century were Joaquim Machado de Assis' romantic novels of the Brazilian people. In the 20th century, the Northeast has produced a particularly wide range of superb writing, including that of Gilberto Freyre on the subject of life under slavery, Graciliano Ramos' tragedies about the drought quadrilateral, and Jorge Amado's light-hearted stories set in the cacao-growing zone of Bahia. Érico Veríssimo's tales of southern Brazil have also been translated into many languages.

Architecture

The landscape architect Roberto Burle Marx has made urban Brazilians especially aware of the splendours of their natural environment by replacing the traditional, formal European-style gardens containing imported plants with a profusion of native species in close approximations of their natural settings. Some of Marx's landscapes have been used to set off the imaginative structures of Brazil's world-renowned architect Oscar Niemeyer. Among his works, Niemeyer designed a striking array of public buildings in Brasília, in collaboration with Lúcio Costa, the creator of the capital's original layout. Brazil also cherishes numerous splendid structures from its colonial and imperial past, from the tiled houses and ornate churches of Salvador to the palaces and public buildings of Rio de Janeiro. Among the most revered of these are the 18th-century churches in Minas Gerais that were adorned by facades, biblical scenes, and statues carved in soapstone by Antônio Francisco Lisboa, better known as Aleijadinho, the "Little Cripple."

Painting

Painting began developing in Brazil in the 18th century, but it was in the 19th century with the works of Belmiro de Almeida, Jr., that a truly original Brazilian art style began to develop, influencing a trend toward realism. In the 20th century the painter Cândido Portinari was influential in developing a uniquely Brazilian style, blending the abstract techniques of Europe with the real people and landscapes of his native land. The classical composer Heitor Villa-Lobos was also a powerful force in breaking with tradition to create distinctively Brazilian compositions by weaving into his music folk themes and rhythms of Portuguese, Indian, and African origins. In contemporary music, João Gilberto introduced the world to bossa nova rhythms, and Sérgio Mendes successfully blended the samba beat into North American and British popular songs.

Not only have the larger Brazilian cities been the scene of contemporary musical extravaganzas, but in Rio de Janeiro, São Paulo, Belo Horizonte, and elsewhere free outdoor concerts of classical music attract multitudes of listeners. Theatres are also widespread and well-attended, from the bawdy regional comedies in neighbourhood theatres to lavish classical productions in Rio de Janeiro's opera house. Art exhibitions range from modest displays in local museums and art galleries, to the prestigious International Biennial of São Paulo; established in 1951, it regularly attracts participants from more than 50 countries. Among the numerous learned societies that have arisen across the country, the Brazilian Academy of Letters, with its headquarters in Rio de Janeiro, is generally regarded as the most prestigious.

Recreation. The most famous Brazilian holiday is celebrated during the four-day, pre-Lenten Carnival. The combination of this Roman Catholic festival with the exuberant celebrations of people of African ancestry has become a lively tradition among the general populace. Carnival evolved principally in the urban centres of the former plantation areas, along the coast between Recife and Rio de Janeiro. In the latter, with its carnival groups, the so-called samba schools (*escolas de samba*), each involving several thousand people, the celebrations have attained their most extravagant expression.

There are various other official and church holidays during the year, including Independence Day, on September 7, and St. John's Night (Noite de São João) in June. The latter is celebrated with bonfires, fireworks, and the launching of small paper hot-air balloons, although the latter are nominally prohibited in the cities because of the fire hazard. Another important festival along the coast is that of the African cult diety Iemanjá, Goddess of the Oceans (also, St. Barbara, patron of seamen); on New Year's Day (a national holiday) fishing vessels head out to sea with offerings that are thought to determine the success or failure of the coming year's catch.

Sports

Class distinctions are not important during celebrations, or among those who attend football (soccer) games, or among those who flock to the country's many beaches. The latter are social gathering spots for young people, where the more athletic play football and various racket games, and for whom municipal governments often provide volleyball courts and other recreational paraphernalia. But football is the nation's most popular sport, and Brazilians are highly enthusiastic fans. It is played virtually everywhere by young and old and amateur and professional, drawing huge crowds to international matches at stadia in the major cities; the largest of these, the Maracanã stadium in Rio de Janeiro, has a capacity of 200,000. Brazilian teams are consistently among the top contenders for the World Cup, and from Brazil came the world-renowned Pelé, one of the greatest players of all time. There are also numerous public parks, both within the cities and at nearby scenic areas, that are used for casual recreation and by family or group picnics.

Parks

In general, state and even national parks are located near urban centres and tend to be heavily provided with amenities. At the other extreme, however, are the newer national parks, which are located in remote areas, particularly in the headwaters of various Amazonian tributaries where they may be adjacent to biologic reserves or Indian reservations. They are not intended for any great number of visitors, and facilities are limited or nonexistent. Park boundaries have been modified when they stand in the way of a new highway or some other development, but since the early 1960s the federal government has become more aware of the need for conservation areas, and a Special Secretariat for the Environment has been established. Among the more popular parks are Itatiaia National Park, Iguaçu National Park, and Serra dos Órgãos National Park.

Press and broadcasting. Brazil's principal newspapers are the *O Estado de São Paulo* and the *Folha de São Paulo,* both in that city, and the *Jornal do Brasil, O Globo,* and *O Dia* in Rio de Janeiro, but numerous others are published in the smaller cities and towns. There are also a number of weekly magazines, including the newsmagazine *Visão* and the glossy pictorial *Manchete.*

Press and broadcasting are intimately linked in Brazil, including television's TV Manchete network and the TV Globo, which, with Rádio Globo, is the largest and most influential of the country's broadcasting media. There are also several lesser networks in radio and television as well as a large number of regional and local stations. An educational network broadcasts to a limited number of major cities. Common television fare includes the tremendously popular prime-time *novelas* (daily hour-long soap operas), sporting events, news, special reports, foreign movies dubbed into Portuguese, and children's programs. In many ways television, in conjunction with massive urban migration, has furthered the homogenization of Brazilian culture and the modification of regional differences.

For statistical data on the land and people of Brazil, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL. (C.W.W./T.L.S./R.P.Mo.)

History

PRE-EUROPEAN BRAZIL

The enormous land area that became Brazil was sparsely populated with indigenous Indians at the time of its discovery by Europeans in the year 1500. Sources have varied in their estimate of the Indian population, although an estimate of more than 2,000,000 is generally recognized. Living for the most part in small tribal groups, the Indians generally were divided into two different groups: the tropical forest peoples and those of marginal cultures. The former group, largely inhabitants of the Amazon

Basin, were somewhat more culturally advanced, having achieved some skill in farming and fishing. The marginal groups lived mainly by hunting and gathering in the drier savanna regions.

The Tupí

Among the more significant of the tropical forest groups were the Tupí, inhabitants of the coastal areas. When the Portuguese arrived, it was the Tupí that they first encountered, and for a considerable period the newcomers carried on their main dealings with that tribal group. Indeed, it may be said that the Tupí were the most important single Indian influence in the early Brazilian colonial period and in the Brazilian culture that developed subsequently. This accomplishment, however, did little to ameliorate the harsh treatment they suffered under European domination.

EARLY PERIOD

Discovery and settlement. The European discovery and early settlement of Brazil appear in the larger perspective of history as episodes in the great movement of Portuguese colonial expansion to the East. Vasco da Gama's discovery of an all-water route to the Indies and the Spice Islands by the Cape of Good Hope in 1498 led the Portuguese crown to dispatch to India an imposing armada under Pedro Álvares Cabral, whose sailing directions had been drawn up by da Gama himself. To avoid the calms off the Gulf of Guinea, Cabral bore so far to the west that on April 22, 1500, the mainland of South America was sighted, and the region was promptly claimed by Portugal, because it lay well within the zone assigned to Portugal by the Treaty of Tordesillas (1494). Portugal's new possession was at first called Vera Cruz (True Cross), a name soon abandoned in favour of Brazil, after the valuable red dyewood (*pau-brasil*) in which the country abounded.

Vespucci's voyage along Brazil's coast

The tidings of Cabral's discovery aroused great enthusiasm, and further explorations were undertaken. One of these expeditions had as its pilot Amerigo Vespucci, whose small fleet sailed along the coast of Brazil and for the first time realized the extent of the land. Vespucci, calendar in hand, baptized the different points on the coast with the names of the saints on whose days they were discovered.

For some years following Vespucci's voyage, interest in Brazil waned. Although a desultory trade with the Indians for brazilwood had begun, the failure to discover precious metals in Brazil focused Portuguese attention on their much more lucrative trade with Asia. Portugal's rivals quickly took advantage of that neglect. The French, in particular, trespassed on Portugal's claims in South America and shipped the valuable red dyewood to Europe. Brazil became a sort of no-man's-land over which the Portuguese crown wielded only a shadowy control. This apathy ended, however, with the accession of John III (reigned 1521–57), and the interest in colonial affairs gradually shifted from Asia to America.

The first systematic effort to establish an organized government in Brazil was made in 1533. Brazil was divided into 15 hereditary captaincies, or fiefs, extending 50 leagues along the coast and an indefinite distance inland. These grants were distributed to favoured persons, chiefly courtiers, who took the names of *donatários*. Their rights and privileges were extensive. The captaincy system failed to realize the hopes of the crown, however, only two of them being truly successful. The most flourishing of the settlements was that of São Vicente, south of the present city of Santos. By the middle of the 16th century this captaincy had a population of about 5,000, including the growing port of Santos and the village of São Paulo on the Serra do Mar's fertile plateau of Piratininga. In northern Brazil, Duarte Coelho Pereira had converted the captaincy of Pernambuco into a great sugar-producing region, offering the first example of a profitable agrarian export from the New World to Europe. To the annoyance of the crown, however, Europeans continued to trespass, threatening Portugal's claims to the New World.

Royal control. King John III determined to substitute for the inefficient rule of the *donatários* a unified, centralized administration. To strengthen royal control, he appointed a governor general for Brazil. Tomé de Sousa, a Portuguese noble with impressive experience in India, took up those duties in 1549, founding a capital for the colony, Bahia (Salvador), from which governors general, later viceroys, governed for the next 214 years. Local officials, responsible to the governor general, were placed over the captaincies, and strategic points along the coast were fortified. In the cities, municipal organizations, similar to those in Portugal, were established. Brazil then began to attract settlers in increasing numbers. By 1600 Bahia and Pernambuco (Recife) each had a population of 2,000 whites, with more than twice as many black African slaves and converted Indians.

Jesuit missionaries

A factor of great importance in the progress of the colony was the labour of the Jesuits. At the request of John III several fathers accompanied Tomé de Sousa to Bahia. Among them was Manuel da Nóbrega, the first of a long line of missionaries who devoted their lives to the protection and conversion of the Indians and to raising the moral level of the colonists. As rapidly as the Indians were converted, they were settled in villages, or *aldeias,* similar to the missions in Spanish America. Jesuit control over a valuable labour supply aroused the ire of the colonists, who did not hesitate to enslave Indians. The conflict between Jesuits and colonists over control of the Indians reverberated throughout the colony. Both parties appealed to the crown. The Jesuits won a partial victory in the royal decree of 1574, granting them full control over the Indians in the *aldeias* while permitting the colonists to enslave Indians captured in legitimate warfare. In the north, the storm centre of a somewhat similar conflict was Father António Vieira, who in the 17th century established a chain of missions in the Amazon Valley. The restrictions on Indian labour stimulated the introduction of black African slaves, a movement that steadily gathered momentum after the middle of the 16th century.

Dutch and French incursion. Brazil had hardly been brought under royal authority before a determined effort was made by the French to establish a permanent colony. In 1555 the French took possession of the beautiful harbour of Rio de Janeiro, which the Portuguese had strangely neglected to occupy. Meanwhile, the Portuguese had come to realize the extent of the French menace. A large force under Mem de Sá, the governor general, blockaded the entrance to the harbour and forced the French garrison to surrender. To ward off future attacks, Mem de Sá in 1567 founded the city of Rio de Janeiro.

From 1580 to 1640 Portugal was united to Spain, and, as a part of the Spanish colonial empire, Brazil was naturally exposed to attacks by Spain's enemies. Among these were the Dutch, who had just succeeded in establishing their independence. They seized and briefly held Bahia in 1624–25, and in 1630 a fleet sent out by the Dutch West India Company captured Pernambuco, the centre of the rich sugar-growing district in the north. As governor of its new possession the company chose Johan Maurits, graaf von Nassau-Siegen, a prince of the House of Orange and perhaps the ablest man in the Netherlands. Distinguished artists and scientists were invited to make known to Europe the resources and beauties of Brazil. But the directors of the company, intent on a policy of greed and gain, refused to support their enlightened governor, and he resigned in 1644. A rebellion launched by a wealthy plantation owner, João Fernandes Vieira, proved beyond the power of Maurits' incompetent successors to crush. The Brazilians themselves defeated and expelled the Dutch in 1654, an achievement that enhanced growing nationalistic sentiments among them.

Westward expansion

Expansion and unification of Brazil. Perhaps the most impressive development of the colonial period was the vast expansion of Brazil beyond the line of Tordesillas (a vertical line drawn 370 leagues west of the Cape Verde Islands)—an expansion that reached westward to the slopes of the Andes and north and south to the Amazon and the Río de la Plata. In the north the movement was led by missionaries, who, in seeking to convert the Indians to Christianity, established missions along the Amazon in the 17th century. In the northeast, cattlemen from the sugar zones of Pernambuco and Bahia pushed inland into the present area of Piauí, Maranhão, and Goiás in search of new pastures. The brunt of the "march to the west," however, was borne by the Paulistas, as the settlers of São Paulo were called. With the purpose of capturing Indian slaves and of finding gold and precious stones, they organized great expeditions into the interior, known as *bandeiras*. Some of the more spectacular *bandeiras* took the Paulistas as far as the silver-mining region of Peru and even to the area of Bogotá in Colombia. In the 17th century they explored the area of Mato Grosso and attacked the Indian *reducciones* (Indian missions) of Misiones and southern Brazil. For the most part they met with resistance from the Indians and their Jesuit protectors; only about the Río de la Plata, in modern Uruguay, did the Spanish settlers resist the invaders, and there eventually the Paulistas were pushed back. The *bandeirante* expansion contributed significantly to the unification of the huge subcontinent of Brazil.

Portuguese as the common language

Other factors added to the growing unification. A unique Luso-Afro-Indian culture emerged. The Portuguese language formed a common bond between plantation residents, cattlemen, miners, slaves (both Indian and black), slave hunters, and city dwellers from the Amazon to the Río de la Plata and distinguished them from their Spanish-speaking counterparts elsewhere in South America. The expanded, patriarchal family structure, also derived from Portugal, was nearly uniform throughout Brazil, and power was exercised in all regions by the heads of those families that controlled the land, slaves, cattle, and, later, mines that produced the wealth of the colony. Despite regional differences, Brazilian society and economy were rurally based. This pattern persisted for a century after the colonial period. The population was divided unevenly between the few who possessed landed wealth and the majority who were dependent upon them. Only three important cities, Bahia, Rio de Janeiro, and Recife, grew up in Brazil during the colonial period. Moreover, contact with Portugal, although at times intermittent, was maintained with all parts of Brazil, and little trade or other regular contact existed between Brazil and neighbouring Spanish colonies. These common factors proved stronger than regional variants when Brazilian unity was under stress in the second quarter of the 19th century.

The colonial economy of Brazil was marked by two major cycles, that of sugar and that of gold. Both reflected the significant role of exports in the Brazilian economy. As mentioned above, the sugar industry, confined primarily to the northeast, was introduced into Brazil early in the 16th century, and it was the principal basis for the wealth of the colony for more than 200 years. Throughout the entire period of Portuguese control in Brazil, sugar produced substantially more revenue for the crown than did any other export product, including gold. A plantation product, sugar called for a large capital investment in land, slaves, and machinery, and consequently it was concentrated in the hands of a relatively small number of wealthy families. Tobacco and cattle raising, conducted chiefly by independent freemen around the periphery of the sugar plantations, became important industries before the end of the colonial period. In the 18th century, cotton and coffee, originally produced by small landholders, were also becoming significant exports.

Discovery of gold in Minas Gerais

Gold, which was sought by colonists in Brazil from the period of first settlement, was not found until 1695 in what is now the state of Minas Gerais. The deposits were the wealthiest that had been discovered to that date, and their discovery set off a gold rush that anticipated in many ways the days of 1849 in California and those of 1851 in Australia. In the hitherto unbroken wilderness, towns sprang up as if by magic, while large sections of the littoral or coastal region were virtually denuded of their inhabitants. African slaves from the sugar plantations and directly from the gold-working regions of Africa introduced many of the techniques employed by Brazilian miners. The yield of the mines during the 18th century has been variously estimated at between £50,000,000 and £200,000,000. The impact of the gold mines upon the Brazilian economy was largely responsible for the transfer of the capital of the colony from Salvador to Rio de Janeiro in 1763. The search for gold led to the discovery of diamonds in the early 18th century in Minas Gerais, Bahia, and Mato Grosso. The gold rush tapered off as the original deposits were depleted, but some gold and diamonds were still being mined in the 20th century.

Colonial reforms. The treaties of Madrid (1750), Pardo (1761), and Ildefonso (1777) with Spain gave some legality to Portuguese claims including the conquests of the *bandeiras*. During this period, Sebastião José de Carvalho e Mello, marquês de Pombal, the prime minister of King Joseph, introduced into Brazil a number of reforms that profoundly affected the social, administrative, and religious life of the colony. The last rights of the *donatários* were absorbed by the crown, the capital was transferred from Bahia to Rio de Janeiro, the Indians were placed on a legal equality with the whites, immigration from the Azores and Madeira was encouraged, two privileged companies were created, and a monopoly of the diamond fields was established. Pombal expelled the Jesuits from Brazil in 1759. Through their protection of the Indians and their various commercial ventures, they had stirred up widespread resentment, and when they attempted to thwart some of Pombal's reforms, the all-powerful minister banished them from both Portugal and Brazil with brutal thoroughness. With an increased centralization of power and by strengthening the lines of command, Pombal progressed toward the unification of Brazil during the final decades of Portuguese rule.

Expulsion of the Jesuits

THE INDEPENDENCE OF BRAZIL

In 1789 the first rebellion against the Portuguese was instigated by José Joaquim da Silva Xavier, who was known as Tiradentes (Tooth Puller) because he occasionally extracted teeth. The uprising was put down, and Tiradentes, who became a national hero, was executed.

Despite its isolation, Brazil could not escape the effects of the revolutionary and Napoleonic epoch in Europe. In 1807 Napoleon invaded Portugal, a traditional British ally, largely to tighten the European blockade of Great Britain. The Portuguese prince regent Dom John, afterward King John VI, decided to take refuge in Brazil, the only time a colony ever served as the seat of government for its mother country. He sailed from the Tagus on Nov. 29, 1807, under the protection of the British fleet. Accompanied by the royal family and a horde of nobles and functionaries, he arrived at Rio de Janeiro on March 7, 1808.

Arrival of Dom John in 1808

The colonists, convinced that a new era had dawned for Brazil, welcomed their sovereign. The Prince Regent promptly decreed a number of reforms that altered the status of the colony. The Portuguese commercial monopoly on Brazilian trade was abolished in 1808, and Brazilian harbours were opened to the commerce of all friendly nations. That same year prohibitions on Brazilian manufacturing were terminated.

To accommodate the imperial administration and to answer local demands, a ministry and a council of state were installed in Rio de Janeiro, and a supreme court of justice, a court of exchequer and royal treasury, the royal mint, the Bank of Brazil, and the royal printing office were established. A royal library, a military academy, and medical and law schools were founded. The status of Brazil was reflected in the decree of Dec. 16, 1815, by which the Portuguese dominions were designated the United Kingdom of Portugal, Brazil, and the Algarves, thus making Brazil coequal with Portugal. In 1816 the queen, Maria I, died, and the Prince Regent became king.

The regency of Dom Pedro

The presence of John VI in Portugal was imperative if the crown was to maintain its authority there. On April 22, 1821, he appointed his son Dom Pedro regent and two days later set sail for Lisbon. Meanwhile, antagonisms between the Portuguese and Brazilians were becoming increasingly bitter, republican propaganda was active, and Dom Pedro had to face the responsibility of a separation of Brazil from Portugal as a result of the shortsighted policy of the Cortes at Lisbon. The majority of this assembly favoured the restoration of Brazil to its former status of colonial dependence. Without waiting for the Brazilian deputies, they proceeded to undo most of the reforms introduced into Brazil by John VI. Fearful that Dom Pedro might head a movement for independence, the Cortes ordered him to return to Europe.

These acts aroused great indignation in Brazil. Supported by the majority of Brazilians, Dom Pedro defied the Cortes by his refusal to return to Lisbon. In January 1822 he formed a ministry in which the chief portfolio was held by the distinguished Paulista, José Bonifácio de Andrada e Silva. Known as the Patriarch of Independence, he was a scholar of European reputation and proved at this critical juncture a tower of strength to the young regent. Events followed rapidly. On June 3 Dom Pedro convoked a legislative and constituent assembly. On September 7, on the plain of Ipiranga, near the city of São Paulo, he solemnly proclaimed the independence of Brazil, and on Dec. 1, 1822, he was crowned the nation's emperor.

The new nation achieved quick recognition, which came first from the United States, in 1824. After the Portuguese acknowledged Brazilian independence in 1825, other European monarchies established diplomatic relations. The relative ease by which Brazil entered nationhood stood in sharp contrast to the experiences of the Spanish-language nations of the New World.

THE BRAZILIAN EMPIRE (1822–89)

The con-
stitution of
Pedro I

The first years of independence were difficult. The nation was without experience in self-government, and the Emperor was inclined to be despotic and arbitrary. When the constituent assembly proved unmanageable, Pedro I dissolved it in 1823 and sent the radical Andrada e Silva and his two brothers into exile. With the aid of the Council of State, Pedro wrote a constitution, considered both liberal and advanced for the period. It provided for the centralization of government, placing strong powers into the hands of the emperor. After the municipal councils debated and approved the document, Pedro promulgated it in 1824. The constitution proved versatile enough to last throughout the imperial period. The popularity of the impulsive Pedro was diminished, however, by a number of occurrences, including the loss of the Cisplatine Province (the present republic of Uruguay) as a result of a costly war with Argentina (1825–28), the humiliating treaties with Great Britain that kept import duties low and exacted a promise to abolish the slave trade, the continuous appointment of Portuguese-born officials to high office, the preoccupation of the Emperor with Portuguese affairs, and the inability of the Emperor to get along with the legislature. As a result, Pedro formally abdicated on April 7, 1831, in favour of the heir apparent, Dom Pedro de Alcântara (Pedro II), then only five years of age.

The regency (1831–40). The next decade proved the most agitated period in Brazilian history. From 1831 to 1835 a triple regency tried in vain to end civil war in the provinces and lawlessness and insubordination in the army. In 1834 the constitution was amended to grant a measure of decentralization to the provinces, through the creation of provincial assemblies with considerable local power, and to provide for a sole regent to be elected for four years. For this office in 1835 a priest, Diogo Antônio Feijó, was chosen. For two years he struggled against disintegration, but he was forced to resign in 1837 and was succeeded by Pedro de Araújo Lima. Sentiment began to crystallize in favour of a declaration of majority of the young Dom Pedro. The Brazilians, impatient with the regency, hoped to find in the Emperor a symbol to which the entire nation might rally. On July 23, 1840, both houses of parliament passed a declaration that Pedro de Alcântara had attained his majority.

Prestige
and
progress
under
Pedro II

Pedro II. The reign of Pedro II, lasting practically a half century, constitutes perhaps the most interesting and fruitful epoch in Brazilian history. The prestige and progress of the nation were due largely to the enlightened statesmanship of its ruler. Though not without personal distinction he was always simple, modest, and democratic. He possessed an insatiable intellectual curiosity and was never happier than when conversing with scholars. He was generous and magnanimous to a fault. One of his favourite occupations was inspecting schools. He was wont to declare, "If I were not emperor I should like to be a schoolteacher." Yet this kindly, genial, and scholarly ruler took his prerogatives and duties as sovereign with great seriousness, and in all matters of first importance he was the final arbiter. According to the moderative power granted to the executive under the constitution of 1824, the emperor had the right to dissolve the Chamber of Deputies, to select the members of the life senate from triple lists submitted by the province, and to appoint and dismiss ministers of state. A kind of parliamentary government functioned under the watchful eye of the Emperor.

Pedro's government took a keen interest in the affairs of the Plata republics, especially of Uruguay, which it sought to control through indirect measures. Brazil aided in the overthrow of the Argentine dictator Juan Manuel de Rosas in 1852. In 1864 Brazil intervened in Uruguayan internal affairs, precipitating war with Paraguay. In alliance with Argentina and Uruguay, Brazil successfully waged the costly and bloody Paraguayan War of 1864–70, eventually overthrowing the Paraguayan dictator Francisco Solano López. This war of the Triple Alliance was the longest and bloodiest war in South American history, and its consequences in Brazil were profound. The empire's relations with the United States and with Europe were generally cordial, Pedro personally visiting Europe in 1871, 1876, and 1888 and the United States in 1876.

Dissension
over
slavery

The empire's major social and economic problems were related to slave-based plantation agriculture. Real political power remained with large rural landholders, who formed the core of the social and economic elite of Brazil. Industrialization was still economically insignificant, and, with the decline of gold mining, agriculture was unrivaled as the source of Brazil's wealth. Cotton, and increasingly after 1840, coffee, competed with sugar as the leading export crop. The rural landholders were largely insulated from the antislavery current of the times. Although manumission was common, and the number of freedmen and their descendants far surpassed the number of slaves in Brazil, the slave owners as a group resisted pressures for the complete abolition of slavery. Partly as a result of pressure from Great Britain, Brazil had agreed to abolish the slave trade in 1831, but it was not until 1851 that slave traffic completely ceased. Agitation to abolish slavery as such began in the 1860s. Pedro was opposed to slavery, but he had to reckon with the slave owners. Yet the ending of slavery internally seemed the next logical step after stemming the importation of slaves. Accordingly, in 1871, the Law of the Free Womb was enacted, declaring free all children born to slaves. The legislation condemned slavery to eventual extinction. But, although agitation subsided briefly, this concession did not satisfy many of the abolitionists. Led by a young lawyer and writer, Joaquim Nabuco de Araújo, they demanded immediate and complete abolition. Nabuco's book *O Abolicionismo* (1883; *Abolitionism*) endeavoured to prove that slavery was poisoning the very life of the nation. In 1884 Ceará and Amazonas freed their slaves; in 1885 all slaves over 60 years of age were liberated. Finally, complete emancipation without compensation to the owners was decreed by the Princess Regent in the absence of the Emperor on May 13, 1888. About 700,000 slaves were freed.

The collapse of the empire. Under Pedro II's wise guidance Brazil had progressed considerably. His rule witnessed a growth in population from 4,000,000 to 14,000,000, a 14-fold increase in public revenues, and a 10-fold increase in the value of the products of the empire. Railroad mileage in 1889 exceeded 5,000 miles; that year more than 100,000 immigrants entered Brazil. Yet there were grave causes of dissatisfaction. The fall of the monarchy has often been ascribed to a restive military, a brooding landed aristocracy, and a resentful clergy. Indeed, those three powerful groups voiced increasingly frequent complaints. Perhaps more pertinent, however, was the widening gulf between a neo-feudal countryside and the more progressive cities and coffee planters during the closing decades of the century, an indication of institutions under severe stress. The urban middle class, the military, and the coffee planters regarded the monarchy as too closely tied to the traditional landed class. The monarchy represented the past; a republic better suited the goals of emerging coffee and industrial capitalism. A conspiracy of those who would propel Brazil toward greater modernization counted on the military to remove the Emperor, and the officers

Abolition of slavery

did so on Nov. 15, 1889. Pedro II abdicated and sailed into a European exile. The abolition of slavery in 1888 and the overthrow of the monarchy in 1889 terminated the two major institutions that had shaped Brazil's past; in so doing they initiated a period of social, economic, and political change that accelerated modernization. Accordingly, the period between 1888 and 1922 witnessed the emergence of a "new Brazil."

BRAZIL SINCE 1889

With support of the military, the nascent middle class, and the prosperous coffee planters, Marshal Deodoro da Fonseca governed Brazil. He established a republic, separated church and state, and, on Feb. 24, 1891, promulgated a new constitution, characterized as presidential, federal, democratic, and republican. The new states of the republic exercised more power than had the empire's provinces.

The "coffee presidents." In 1891 Congress elected Deodoro da Fonseca as president, but he proved unable to govern under the new constitution. His act of forcibly dismissing a dissenting Congress raised such public outcry that he resigned. The popular vice president, Marshal Floriano Peixoto, ascended to office on Nov. 23, 1891. By defeating several monarchist and military revolts, he restored tranquillity and order to the nation. In 1894 he turned over the presidency to the first civilian president, Prudente de Morais, the first republican governor of coffee-rich São Paulo. Under the successive "coffee presidents," primarily from the states of São Paulo or Minas Gerais, Brazil enjoyed peace, financial reform, and mounting coffee exports.

Growing dependence on coffee exports

The economic and political centres of the nation had shifted from the old sugar regions of the Northeast to the new coffee regions of the Southeast. In an economy overly dependent on exports for its vitality, coffee dominated, accounting for 46 percent of the exports by 1901 and 53 percent by 1908. Everything went smoothly for the coffee planters and the national economy until the early years of the 20th century, when overproduction threatened coffee—and, thus, national—prosperity. To counter that threat, representatives of the three major coffee-producing states, São Paulo, Minas Gerais, and Rio de Janeiro, signed a federal government-backed agreement at Taubaté in 1906 to guarantee a fair world market price for coffee. The agreement inaugurated the scheme of coffee valorization—the purchase of excess coffee and its removal from the market in order to maintain its price.

The trends of immigration and urbanization accelerated. The rubber boom in the distant Amazon Basin changed isolated Manaus into a cosmopolitan city with electricity, streetcars, cinemas, a grandiose opera house, and an international society. Under the direction of prefect Francisco Pereira Passos, Rio de Janeiro was transformed into one of the world's most beautiful cities, and also a more healthful one, due to the eradication of yellow fever by the distinguished physician and scientist Oswaldo Cruz. Coffee prosperity metamorphosed São Paulo, whose population jumped from 35,000 in 1883 to 350,000 in 1907, into the bustling commercial centre of the nation. Santos became one of the world's busiest ports, sending coffee exports to the cities of Europe and North America.

During the time that great material change was taking place, Brazil was also undergoing a literary renaissance. The intellectuals at the turn of the century fixed their eyes on the new Brazil and analyzed it carefully. Two of the foremost novelists, Joaquim Maria Machado de Assis and Afonso Henriques de Lima Barreto, took a penetrating look into Brazilian society. In his masterpiece, *Os Sertões* (1902; *Rebellion in the Backlands*), Euclides da Cunha described the clash between the backlands with their timeless tradition and the cities as the focus of modernity: the conflict between the "two Brazils."

Territorial gains of Rio Branco

Under the brilliant leadership of the barão do Rio Branco, José Maria da Silva Paranhos, a diplomat who became foreign minister, the Luso-Brazilian policy of territorial expansion reached its fruition between 1895 and 1909. Using the principle in international law of uti possidetis, which grants territory to the power that absolutely controls it, he advantageously closed the 9,000 miles of borders separating Brazil from its neighbours and won Brazil's claims to approximately 342,000 square miles of territory, an area larger than France. Somewhat more difficult to measure, but of greater importance, was his elimination of the potential causes for war, misunderstanding, and dispute. As foreign minister in 1902–12, Rio Branco shifted Brazil's diplomatic axis from London to Washington, D.C. Sympathetic to the Allied cause, Brazil declared war on Germany on Oct. 26, 1917. Brazil participated in the peace conference and held a temporary seat on the Council of the League of Nations.

Rebellion against the coffee elite. Resentful of the political and economic institutions favourable to the coffee planters, the growing urban middle class agitated for changes. So did the military, particularly the junior officers. In the elections of 1922, urban groups and the military united in an unsuccessful effort to wrest the presidency from the coffee elite. Thereafter, violent protest erupted in Brazil. In July of 1922, shortly after the coffee interests had once again imposed their presidential preference on the nation, a handful of junior officers revolted in Rio de Janeiro in a poorly planned effort to overthrow the moribund republic. It signaled the beginning of an eight-year period of unrest that came to a climax in 1930 with the fall of the old republic. Military uprisings of a more threatening nature—once again dominated by junior officers—broke out in 1924. Survivors of that rebellion marched thousands of miles through the interior in the mid-1920s attempting to stir up revolt. The countryfolk were still too dominated by the local landowners to be enlisted into the ranks of rebellion. The countryside never had been the place to foment change in Brazil, and the failure of the efforts in the mid-1920s proved that it still was not.

It was in urban Brazil that the demand for social and political progress grew. Sharpened by cultural nationalism, especially promoted by the Modern Art Week in São Paulo in 1922, criticism of the "coffee governments" mounted within the cities. The nationalists denounced the selfish regional monopolization of power, manipulated elections, and continuation of colonialism.

The *tenentes*

By 1926 the movement of the junior military officers, known as *tenentes* (meaning "lieutenants"), had acquired a somewhat more identifiable, though imprecise, nationalistic philosophy of political and economic development. The *tenentes* maintained a mystical faith that somehow the military could alter the habits of the country and provide the impetus to propel it into the modern age. Their primary concern was not democracy but reform and development. They wanted to retire the entrenched politicians and begin the modernization process; then would they consent to return the nation to constitutional rule. They hoped to expand the base of the government and to eradicate regionalism, while favouring a strong, centralized government. Revealing social democratic tendencies, the *tenentes* proposed government recognition of trade unions and cooperatives, a minimum wage, maximum working hours, child-labour legislation, agrarian reform, nationalization of natural resources, and expanded educational opportunities. Obviously much of the program advocated by the *tenentes* favoured the goals of the urban middle-class groups. Those groups, however, failed to understand that the military rebellions and movements in the 1920s could have been turned to their advantage; they were unable to coordinate their desires for modernization with the similar desires of the young officers. It was, finally, two related events that ended the monopoly of political power held by the coffee elites. The first was the precipitous decline of coffee prices caused by the international financial crisis of 1929–30; the second was the determination of the politicians of São Paulo to impose yet another one of their group as president of the nation in 1930. These movements and events combined to bring to an end the old republic and mark the beginning of a new era.

The Vargas era. When the administration-sponsored presidential candidate was officially declared winner in the 1930 election, a revolt broke out. This successful rebellion was led by Getúlio Vargas, governor of the state of Rio Grande do Sul, who was to remain central in Brazilian national life until his suicide in 1954.

The Estado Nôvo

Vargas held office as chief executive on two occasions, 1930–45 and 1951–54. Early in the first of these periods, the country's difficulties were essentially economic, accentuated by a policy of subsidizing coffee production, and the tremendous powers of the states served as steady irritants to the country. In 1934 a new constitution gave the central government greater powers. Three years later, President Vargas seized practically absolute power and set up still another constitution, under which he continued as president. The new administration, known as the Estado Nôvo (New State), so concentrated power in Vargas' hands that he was able to suppress all manifestations of popular will. Vargas stripped Brazil of most of the trappings through which it might eventually hope to become a democracy. However, social legislation was enacted, with positive benefits to the labouring class.

Vargas' policy concentrated upon curtailment of states' rights, emphasis on nationalism, and the transfer of the locus of power from the country to the city. The states were increasingly subordinated to the central government, politically, economically, and socially. His government urged greater diversification of agriculture and the further development of industrialization, whose import-substitution phase accelerated impressively.

After the outbreak of World War II in 1939, Vargas' government supported inter-American solidarity; on Aug. 22, 1942, it declared war against Germany and Italy. Besides participating in the defense of the South Atlantic, Brazil sent an expeditionary force to Italy in July 1944 that distinguished itself in several battles. Brazil also allowed the United States to use certain Brazilian naval and air bases during the war. Agreements were signed between the two governments for the economic development and production of Brazilian raw materials.

On Oct. 29, 1945, the military forced President Vargas to resign as it began to appear more likely that he would attempt to retain power. From his resignation there initiated a period of experimentation with democracy.

The democratic interlude. General Eurico Gaspar Dutra, Vargas' own choice, won the presidential election in December 1945. Vargas himself was elected to a seat in the Senate. In September 1946 a new constitution was promulgated, Brazil's fifth and the fourth of the republican era. In reaction to the Vargas experience, the new constitution sought to prevent the rise of another overpowering president. While the presidency remained the key institution, the office was sufficiently circumscribed to inhibit a future leader from abuse of power. The new constitution separated the three branches of government, ensuring Congress of its independence and freedom in the election of its members. Restrictions were put in place to prevent abusive federal intervention in the internal affairs of the states.

Return of Vargas to power

The general elections of 1950 returned Vargas to power, however, by a substantial margin. While he failed to secure a clear majority in the four-way race, he polled 1,500,000 more votes than the runner-up and nearly as many as the combined total for the three rival candidates. Accordingly, despite serious reservations by the military leaders who had deposed him in 1945, he was installed in the presidency once more on Jan. 31, 1951. During his second period in power, Vargas was unable to dominate the political forces of the country or to exploit social and economic trends to his advantage; he endeavoured to abide by the constitution of 1946 and was criticized in some quarters for weak leadership. Lacking a firm majority in the Congress, he could neither enact his own program nor resist the contradictory pressures of his supporters and opponents. The most serious continuing problem was economic. Brazilians agreed upon the need for more rapid industrial development and measures to ameliorate the effects of gradual monetary inflation, which persisted as government expenditures consistently outran revenues. Vargas maintained a precarious balance between the advocates of greater state intervention in the economy and those who insisted upon a favourable climate for domestic and foreign private investment. The situation was complicated by the campaign to nationalize the natural resources of the country. The campaign culminated in the creation of the national petroleum corporation, Petrobrás, in 1953, a formidable nationalist victory.

Vargas' suicide

For three years Vargas' popularity largely protected him from attack by political adversaries, who directed their criticism against members of his administration. João Goulart, Vargas' young protégé and vice president of the Brazilian Labour Party (Partido Trabalhista Brasileiro; PTB), who became minister of labour in 1953, was accused of using his office to transform organized labour into a political machine loyal to Vargas. His sponsorship of a decree doubling the minimum wage, thus strengthening labour–administration ties and contributing to the inflationary spiral, led to his dismissal in 1954. A series of crises that followed reached a climax on Aug. 5, 1954, by the murder of an air force officer and the attempted assassination of opposition newspaper editor Carlos Lacerda. Investigation revealed that the assassins had been hired by the President's personal guard. The exposure of widespread corruption within the administration caused a wave of antipathy for the former dictator. A group of army officers presented Vargas with a demand for his resignation; Vargas complied on Aug. 24, 1954, and later that day committed suicide.

Kubitschek's administration. Vice President João Café Filho served out the remainder of Vargas' term. Much of the Café Filho administration was occupied by preparations for the presidential election of Oct. 3, 1955. After the failure of attempts to find a single candidate around whom all major political parties might rally, three strong candidates for the presidency emerged: former Minas Gerais state governor Juscelino Kubitschek de Oliveira, popularly regarded as Vargas' political heir; former São Paulo state governor Ademar de Barros, with broad backing from financial and commercial groups; and Marshal Juárez Távora, considered as the representative of conservative military and civilian groups. In the 1955 election Kubitschek was the victor with slightly more than one-third of the total vote; in the separate vice-presidential election, João Goulart won with a large plurality. The election was widely interpreted as a popular vindication of the Vargas position. The illegal but active Communist Party, which had thrown its unsolicited support to Kubitschek, claimed to have provided his margin of victory. Rumours of a Café Filho administration plot to use unconstitutional means to prevent the inauguration of Kubitschek and Goulart were not proved, but the conservative press—which regarded Kubitschek as a dangerous radical—added to the tension. President Café Filho suffered a mild heart attack and on Nov. 8, 1955, transferred power temporarily to his constitutional successor, Carlos Luz, speaker of the Chamber of Deputies. The following day Luz indicated that he would dismiss the war minister Marshal Teixeira Lott to resolve a conflict of authority between the executive and the war minister. This was interpreted by Kubitschek's backers as the signal for a coup d'état by anti-Kubitschek forces. On Nov. 11, 1955, Teixeira Lott and Marshal Odílio Denys, commander of army troops in Rio de Janeiro, staged a "countercoup" to guarantee Kubitschek's inauguration. Luz was deposed, and Kubitschek and Goulart took office as scheduled and without incident on Jan. 31, 1956.

Spread of nationalism

During his term of office Kubitschek encouraged the widespread spirit of nationalism, appealing to the popular demand for economic development and to the belief that Brazil was destined to become a great power among the nations of the world. Kubitschek felt that the national government should play a vital role in economic areas unattractive to private capital; thus, his administration undertook ambitious developmental programs for the construction of highways and hydroelectric power projects, the expansion of iron, steel, petroleum, and coal production, and assistance to the private industrial sector. Perhaps his outstanding accomplishment was the construction of Brasília 580 miles northwest of Rio de Janeiro. Brasília was dedicated as the official capital of Brazil on April 21, 1960. The purpose of Brasília was to focus attention on the interior of the country and to hasten settlement of the region and the development of its untapped resources. The new capital was denounced by residents of Rio de Janeiro, but Brazilians elsewhere generally regarded it as a symbol of the nation's future greatness. In inter-American relations, the Kubitschek administration proposed the adoption of Operation Pan America, a cooperative program for the economic development of Latin America, which foreshadowed the Alliance for Progress.

The great material progress during the Kubitschek period was achieved at a high cost in terms of inflation and repeated foreign borrowing. The cost of living and the volume of currency in circulation tripled between 1956 and 1961, while Brazil's large foreign debt nearly doubled. The gross national product rose to unprecedented levels, but living standards for much of the population remained stationary or actually declined. At the same time, indications of large-scale graft and favouritism in public office disturbed even the normally tolerant Brazilian electorate.

Political turmoil. In the hotly contested elections of 1960, the National Democratic Union (União Democrática Nacional; UDN), the largest conservative party, nominated Jânio Quadros, a maverick politician who had governed São Paulo successfully. He triumphed at the polls. At the same time, João Goulart, the vice-presidential candidate of Vargas' PTB and heir to the political mantle of Vargas, won. Those two politically divergent politicians took office on Jan. 31, 1961.

Control by anti-Vargas groups

The election of Quadros was hailed as a revolution by ballot. For the first time in three decades, anti-Vargas political groups controlled the presidency. Quadros took office in an atmosphere of popular expectation, but almost immediately he encountered resistance from the Congress, where parties loyal to the Vargas tradition still commanded a large majority. Quadros' arbitrary and autocratic manner, which soon alienated many of his former adherents, further reduced the possibility of enacting a meaningful reform program or effective measures to retard inflation. In foreign affairs Quadros was more successful; his foreign policy, which was applauded by ultranationalists and deplored by moderates, seemed designed to move Brazil toward neutralist nations and away from the traditional ties with the United States. He opposed any inter-American action to censure the Castro regime in Cuba and initiated proceedings to resume or establish diplomatic and commercial relations with the Soviet Union and its European satellites. On Aug. 25, 1961, after less than seven months in office, Quadros resigned unexpectedly, alleging that "terrible forces" had deterred him. The Congress promptly installed Pascoal Ranieri Mazzilli, speaker of the Chamber of Deputies, as temporary president. Vice President Goulart, the constitutional successor, was then en route home from a state visit to China.

Succession crisis

Brazil stood at the brink of civil war. Many military commanders and conservatives regarded Goulart as too radical to be entrusted with the nation's highest office, while the great majority of civilian political leaders upheld his constitutional right to the presidency. War Minister Odílio Denys now emerged as the chief spokesman of the anti-Goulart forces, demanding that the Congress declare the office of vice president vacant and hold new elections. The Congress refused. In southern Brazil the commanders of the powerful army and air force units defied orders from the capital and sided with Goulart. Goulart arrived in Pôrto Alegre, insisting that he was already president of Brazil. Faced with the prospect of armed conflict, the Congress and the anti-Goulart group in the military agreed upon a compromise solution: that Goulart be allowed to take office, but only as a figurehead. On Sept. 2, 1961, the parliamentary system of government was adopted, and most presidential powers were transferred to the newly created post of prime minister. Provision was made for a national plebiscite on the parliamentary experiment, and Goulart was confirmed as president.

After the congressional elections of October 1962 failed to indicate a clear mandate for or against the administration's reform proposals, the opponents of parliamentarianism, led by Goulart, demanded a quick return to presidential rule. On Jan. 6, 1963, a national plebiscite by a margin of more than five to one gave Goulart full presidential powers. Goulart was unable to secure majority support for his legislative program, however, and new plans for economic and social development did nothing to restrict inflation, which reached alarming proportions. During Goulart's administration, there was a 10-fold drop in the dollar value of the currency, the cost of living tripled, and the growth of the gross national product, which had been rising by 6 to 7 percent yearly, was brought to a complete halt.

Opposition to Goulart

Military intervention. As the situation grew more desperate, cooperation between the regime and its critics became more difficult. The administration identified itself increasingly with the ultranationalistic left, while most of the military officer corps sympathized with the moderate and conservative opposition. Goulart surrounded himself with left-wing advisers. He sought to neutralize the armed forces by frequent command changes and by developing a personal following among noncommissioned officers and enlisted personnel. After failing to obtain authorization for a state of siege, which would have enhanced his powers, Goulart appealed directly to the people for support. At a mass rally in Rio de Janeiro in March 1964, he instituted a controversial agrarian reform and nationalized certain privately owned oil refineries. Many military officers and opposition political leaders had already become convinced that Goulart was seeking to establish a dictatorship of the left. During 1963 active plotting began among separate civilian and military groups in different parts of the country. Governor José de Magalhães Pinto of Minas Gerais state and Marshal Humberto de Alencar Castelo Branco, chief of staff of the army, emerged as the chief coordinators of the conspiracy. The administration's refusal to suppress a strike by naval enlisted men in late March 1964 was regarded by the opposition as intolerable, for military authority and discipline were considered the last check on Goulart's alleged ambitions. On March 31, 1964, Magalhães Pinto proclaimed a rebellion against the government by the civil and military forces in Minas Gerais; he was joined by key political leaders and by most of the armed forces. On April 2 Goulart fled into exile and the Congress declared his office vacant; Ranieri Mazzilli was again designated as interim president.

The First Institutional Act

Military dictatorship. With the fall of Goulart, real power passed to the leaders of the rebellion, who instituted sweeping political changes. The commanders set out to restore economic and financial order, to eliminate what the generals claimed was "Communist infiltration," and to purge corrupt and subversive elements; they desired to retain a modified form of representative government. On April 9, 1964, these goals were combined in the First Institutional Act, a modification of the 1946 constitution. The executive was granted temporary authority to cancel the mandates of elected officials, to dismiss public servants, and to revoke for 10 years the political rights of those found guilty of subversion or misuse of public funds.

During the six months these provisions were in effect, thousands were arrested, and hundreds of persons—including union and government officials and former presidents Goulart, Quadros, and Kubitschek—were deprived of political rights. The Congress was retained as a legislative body with power to debate and to amend but not to reject proposals submitted to it by the executive. On April 11, 1964, Castelo Branco was elected president by the Congress. (L.W.B./A.S.G./G.I.B./R.E.P./E.B.B.)

Castelo Branco's extended term was regarded by the military as a transition period during which sweeping political and economic reforms should be enacted before the nation could again be entrusted to a popularly elected government. The new leaders agreed on goals, such as the curtailment of multiple political parties, but they were divided over the best means to attain their ends. The President's determination to achieve reform by legislation, while permitting normal political activities, was challenged by civilian and military extremists, who called for dissolution of Congress, abolition of political parties, and suspension of all partisan activities until the military regime could consolidate its power.

The test came in October 1965, when gubernatorial elections were held in 11 states, and opposition candidates were elected by substantial majorities in the key states of Minas Gerais and Guanabara. The extremists interpreted the results as a great setback for the government, and they demanded that Castelo Branco annul the two elections they had lost. When he refused to comply with these demands, they plotted a coup d'état. At this point, Marshal Artur da Costa e Silva, the war minister, intervened and persuaded the dissident leaders to keep the peace in return for Castelo Branco's promise that he would embrace the military's reform program.

On Oct. 27, 1965, Castelo Branco signed the Second Institutional Act, which suspended all existing political parties and restored the president's emergency powers for the remainder of his term, and set Oct. 3, 1966, as the date for new presidential elections. An artificial two-party system was created. The new political parties were the National Renewal Alliance (Aliança Renovadora Nacional, called ARENA), sponsored by the government, and an opposition party, the Brazilian Democratic Movement (Movimento Democrático Brasileiro; MDB). ARENA attracted a grouping of pro-administration elements, while the MDB came to be composed largely of politicians who had broken with the military regime. Because the president was elected by the ARENA-dominated Congress, the MDB refused to nominate candidates. The administration candidate, Costa e Silva was elected and was scheduled to replace Castelo Branco in 1967.

Castelo Branco in December 1966 called an extraordinary session of Congress to approve a new constitution drafted by a government-appointed commission. The new document, promulgated on Jan. 24, 1967, incorporated much of the military's program and confirmed the expansion of the power of the executive and of the central government. Important changes included the indirect election of the president and vice president from a single ticket, reduction of the presidential term from five to four years, the use of military courts to judge civilians charged with violation of national security laws, the granting of authority to the president to issue emergency decrees without consulting Congress, and denial to the Congress of the right to delay action on legislation requested by the executive.

Constitution of 1967

The Castelo Branco administration engaged in unprecedented economic planning in order to contain inflation and to revive the flagging economy. It employed emergency powers to limit and regulate sources of credit, to restructure the tax system and collection procedures, and to impose wage and salary controls. At the same time the government poured investments into electric power and transportation to provide for continued orderly economic development. Although the administration did not achieve all of its goals, Brazil's credit rating abroad was firmly reestablished, the annual rise in living costs fell from 86 percent in 1964 to 41 percent in 1966, and the increase in the gross national product rose significantly. With the significant exception of labour, every major sector of the economy was expanding when Castelo Branco left office.

Administrations of Costa e Silva, Médici, and Geisel. Although Costa e Silva promised to humanize the military government, he did not depart markedly from the course set by his predecessor. His administration rejected petitions for a general amnesty, resisted proposals to amend the new constitution to restore direct elections, quashed attempts to form a second opposition party, and suppressed student disturbances. More serious political opposition, however, was muted in part by the government's economic achievements.

The political situation deteriorated rapidly late in 1968. Faced with a resurgence of public and congressional criticism, Costa e Silva seized emergency powers. The Fifth Institutional Act issued on December 13 suspended all legislative bodies indefinitely, authorized the executive to rule by decree, and provided the legal basis for a new purge of outstanding political critics. (R.E.P./E.B.B.)

Growing opposition

In August 1969 Costa e Silva suffered a stroke, and the government was run by the three armed forces ministers until October, when the government selected General Emílio Garrastazú Médici as the new president. Elections for federal and state senators and deputies and municipal officials in November 1970 showed a clear victory for Médici and the ARENA party. Disturbances continued through 1970 and 1971, kidnapping of foreign consular and diplomatic officials being added to the arsenal of the dissidents.

Médici in 1971 presented the First National Development Plan, which was aimed at 8 to 10 percent annual economic growth and development of the Northeast and the Amazon area, especially by means of road construction and redistribution of land. An electoral college was created in 1973, and on Jan. 15, 1974, the ARENA party's General Ernesto Geisel was elected president.

The 10th anniversary of the military coup d'état was celebrated by lifting the prohibition on political activities of 106 leaders of the former regime, among them Kubitschek, Quadros, and Goulart. The Fifth Institutional Act, however, remained in force. A surprising victory for the MDB in the congressional elections in November 1974, in which the opposition gained 20 seats in the Senate, was not repeated in the municipal elections of November 1976.

In April 1977 President Geisel dismissed Congress when it failed to pass judicial reforms that he had requested. He then used the emergency powers of the Fifth Institutional Act to institute those reforms and other electoral and constitutional changes. Among the reforms were provisions for the indirect election of state governors and one-third of the federal senators and the increase of the presidential term to six years. The number of members of the Chamber of Deputies was to be based on the total population of the states instead of on the number of registered voters, and constitutional amendment could be effected by an absolute majority of Congress rather than the two-thirds vote of two successive sessions formerly required.

Reforms under Geisel

Transition to democracy. In November 1978 Geisel's handpicked successor, General João Baptista de Oliveira Figueiredo, won the election for president. Before leaving office, Geisel repealed all emergency legislation, including the presidential decree (of 1969) providing for the banishment of persons accused of political crimes. (The Fifth Institutional Act had been repealed by constitutional amendment in October 1978.) In 1979 Congress enacted an amnesty, restoring political rights to all who had lost them since 1961. Political freedom returned, and a reinvigorated freedom of expression sparked lively political debate. Deeper transformations also characterized Brazil. It had become one of the major industrial nations of the world, boasting the world's 10th largest gross national product by the early 1980s. At the same time, fully 70 percent of the population was urban. The transportation infrastructure had expanded impressively, and road networks in particular reached out to previously isolated corners of the vast nation. New groups played increasingly influential roles, notably organized labour, and the social structure had become more widely diverse and complex.

Still, Brazil followed in the 1980s well-delineated patterns of the past. The few still governed the many and enjoyed most of the benefits of society. The large estates grew in size and number as Brazil's agricultural frontier moved ever westward and through the Amazon. The export sector still dominated and shaped the economy. Poverty characterized the lives of the overwhelming majority of Brazilians. Indeed, Brazil did not escape the economic crises shaking Latin America in the 1980s. Its foreign debt ranked as the largest in the Third World. The nation emerged from the period of military dictatorship with a triple-figure inflation. Nor had the military governments resolved the problems of illiteracy, malnutrition, and high infant mortality that plagued the majority of the people.

Return of civilian government. In an indirect election in January 1985, the electoral college selected the candidates of the MDB, Tancredo de Almeida Neves for president and José Sarney for vice president, thus repudiating the candidates favoured by the military. Before Neves could assume office in mid-March 1985, he died from the effects of an illness. Subsequently, Vice President Sarney was inaugurated as Brazil's first civilian president since 1964. The period of military dictatorship had ended, and Sarney presided over the inauguration of the "new republic" as a constituent assembly prepared a new constitution. The political crisis of redemocratization; the economic crises of debt, inflation, recession, and unemployment; and the social crises of poverty and injustice confronted him and, in a larger sense, challenged the nascent democracy.

The Sarney administration

After Sarney took office rapid economic expansion took place as agricultural production rose and the country reacted favourably to new economic and political policies. Progressive steps taken by the government included a measure that would provide land to several million peasants by the year 2000 and new legislation that legalized all political parties and paved the way for direct presidential elections. Sarney's approval rating ran high as his government imposed the Cruzado Plan, an anti-inflationary control package including wage and price freezes that further fueled the economy. By the end of 1986, however, the government allowed price increases to brake the overheated economy. Inflation immediately began to rise, precipitating massive protests against the government. The crisis took place right after a new congress had been elected (November 1986) and endowed with the task of producing a new constitution.

The Constituent Assembly began its deliberations in February 1987 as the failed Cruzado Plan ended. A year and a half later, on Oct. 5, 1988, Brazil's eighth constitution was promulgated. The document provided for many new public freedoms, such as freedom of assembly and thought, the right of workers to strike and engage in collective bargaining, and the abolition of government censorship of art and literature. The principle of habeas corpus was introduced, and the voting age was lowered to 16. One of the most hotly debated questions of the assembly had been the length of the presidential mandate, which was finally set at five years. The constitution also provided for a presidential election in November 1989 and prohibited the president from enacting laws by decree. A plebiscite was to be held in 1993 on whether to keep Brazil's presidential system of government or to change to a parliamentary form or a monarchy. (E.B.B./Ed.)

World concern for the rain forests

World attention was increasingly drawn to Brazil by the continued destruction of the Amazon rain forests by cattle ranchers, miners, lumber companies, and rubber companies, especially after the murder in December 1988 of the internationally known labour leader and ecologist Francisco "Chico" Mendes Filho. In response to international pressure, Sarney created a 2,000-member National Environment Guard and promised to use the armed forces as well as the federal police to deter further deforestation.

The presidential elections dominated 1989. Early favourites and major party leaders Leonel Brízola, of the Democratic Labour Party, and Ulysses Guimarães, head of the MDB Party, the largest party, did not do well in the November elections. Instead, the race was essentially between Luís Inácio da Silva ("Lula"), leader of the Workers' Party, and Fernando Collor de Mello of the newly formed National Reconstruction Party. No candidate received a majority of votes, however, and a runoff between Lula and Collor was held in December. Collor, a former governor of the northeastern state of Alagoas and a centrist, won the election with about 43 percent of the vote. Lula received 38 percent, mostly from the larger cities. The remaining ballots were disallowed or left blank.

Collor, who took office without a parliamentary majority, faced severe problems with the Brazilian economy. Immediately upon taking office on March 15, 1990, he announced a series of measures to eliminate inflation, control the enormous debt, and encourage economic growth. Collor saw excessive government spending, the practice of income indexing, widespread financial speculation, and trade barriers as the root causes of the country's economic problems. (Ed.)

For later developments in the history of Brazil, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 964, 966, and 974, and the *Index.*

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Broadcasting

Broadcasting is the transmission of radio and television programs that are intended for general public reception, as distinguished from private signals that are directed to specific receivers. In its most common form, broadcasting may be described as the systematic dissemination of entertainment, information, educational programming, and other features for simultaneous reception by a scattered audience with appropriate receiving apparatus. Broadcasts may be audible only, as in radio, or visual or a combination of both, as in television. Sound broadcasting in this sense may be said to have started about 1920, while television broadcasting began in the 1930s. With the advent of cable television in the early 1950s and the use of satellites for broadcasting beginning in the early 1960s, television reception improved and the number of programs receivable increased dramatically.

The scope of this article encompasses both the technical and the nontechnical aspects of broadcasting. It traces the development of radio and television broadcasting, surveys the state of broadcasting in various countries throughout the world, and discusses the relationship of the broadcaster to government and the public. Discussion of broadcasting as a medium of art includes a description of borrowings from other media. In addition, the article highlights important developments in radio and television technology and delineates principal systems and equipment. For more detailed information about electronic components and techniques used in radio and television communications, see the articles ELECTRONICS and TELECOMMUNICATIONS SYSTEMS. (J.A.Ca./Ed.)

For coverage of related topics in the *Macropædia* and the *Micropædia*, see the *Propædia*, sections 735 and 10/32, and the *Index*.

The article is divided into the following sections:

History

RADIOBROADCASTING

The United States. The first known radio program in the United States was broadcast by Reginald Aubrey Fessenden from his experimental station at Brant Rock, Massachusetts, on Christmas Eve, 1906. Two musical selections, the reading of a poem, and a short talk apparently constituted the program, which was heard by ship wireless operators within a radius of several hundred miles. Following the relaxation of military restrictions on radio at the conclusion of World War I, many experimental radio stations—often equipped with homemade apparatus—were operated by amateurs. The range of such broadcasts was only a few miles, and the receiving apparatus necessary to hear them was mostly in the hands of other experimenters, who, like the broadcasters, pursued radio as a hobby. Among the leading personalities of this early period was David Sarnoff, later of the Radio Corporation of America and the National Broadcasting Company, who first, in 1916, envisaged the possibility of a radio receiver in every home.

Growth of commercial radio. From this beginning the evolution of broadcasting was rapid; many persons who wanted to hear music from the air soon created a demand for receivers that were suitable for operation by the layman. The increase in the number of listeners in turn justified the establishment of stations especially for the purpose of broadcasting entertainment and information programs. The first commercial radio station was KDKA

First commercial radio station

in Pittsburgh, which went on the air in the evening of November 2, 1920, with a broadcast of the returns of the Harding–Cox presidential election. The success of the KDKA broadcast and of the musical programs that were initiated thereafter motivated others to install similar stations; a total of eight were operating in the United States by the end of 1921.

The popularity of these early stations created two possible sources of financial support to offset the operating costs of broadcasting. First, there were possibilities for profit in the manufacture and sale of radio receiving equipment; and, second, the fame attained by the organizations operating the first broadcasting stations called attention to the value of broadcasting as an advertising medium. Advertising eventually became the principal means of support for broadcasting in the United States.

Between 1921 and 1922 the sale of radio receiving sets and of component parts for use in home construction of such sets began a boom that was followed immediately by a large increase in the number of transmitting stations. By November 1, 1922, 564 broadcasting stations had been licensed.

Interconnection of stations. The use of long-distance wire telephone lines in 1922 to connect a radio station in New York City with one in Chicago to broadcast a description of a football game introduced a new idea into radiobroadcasting. In 1926 the National Broadcasting Company purchased WEAF in New York and, using it as the originating station, established a permanent network of radio stations to which it distributed daily programs. Some of these programs were sponsored by advertisers and furnished revenue to both the network and its associated stations, while others were supported by the network, with a portion of the time being set aside for public-service features.

Government regulation. Although the growth of radiobroadcasting in the United States was spectacularly swift, in the early years it also proved to be chaotic, unplanned, and unregulated. Furthermore, business arrangements that were being made between the leading manufacturers of radio equipment and the leading broadcasters seemed to threaten monopoly. Congress responded by passing the Radio Act of 1927, which, although directed primarily against monopoly, also set up the agency that is now called the Federal Communications Commission (FCC) to allocate wavelengths to broadcasters. The government's attack on monopoly resulted eventually in four radio networks—the National Broadcasting Company, Columbia Broadcasting System, Mutual Broadcasting System, and American Broadcasting Company—while the FCC permitted orderly growth and assured the survival of educational radio stations.

First trans-Atlantic broadcast

The United Kingdom. *Early development.* Radiobroadcasting in Great Britain eventually developed in quite a different way from that in the United States. The first initiatives after World War I were taken by commercial firms that regarded broadcasting primarily as a means of point-to-point communication. The first successful broadcasting of the human voice, from a transmitter in Ireland across the Atlantic in 1919, led to the erection of a six-kilowatt transmitter at Chelmsford, Essex. From this spot two daily half-hour programs of speech and music, including a well-received broadcast by the opera singer Dame Nellie Melba, were broadcast for about a year between 1919 and 1920. Opposition from the armed services, fear of interference with essential communications, and a desire to avoid the "commercialization" of radio led, however, to a ban on the Chelmsford broadcasts, which the Post Office claimed the right to impose. Experimental broadcasts, the Post Office ruled, had to be individually authorized. Nevertheless, about 4,000 receiving-set licenses and 150 amateur transmitting licenses issued by the Post Office by March 1921 were evidence of growing interest. When these amateurs, grouped into 63 societies with a total of about 3,000 members, petitioned for regular broadcasts, their request was granted in a limited form: the Marconi Company was authorized to broadcast about 15 minutes weekly.

The first of these authorized broadcasts, from a hut at Writtle, close to Chelmsford, took place on February 14, 1922; the station call signal was 2MT. Shortly thereafter an experimental station was authorized at Marconi House in London, and its first program went on the air May 11, 1922. Other stations were soon to follow.

Revenue sources of the British Broadcasting Company, Ltd.

Formation of the British Broadcasting Company. By this time developments in the United States had demonstrated the commercial possibilities of radio but also suggested a need for greater order and control. The Post Office took the initiative in encouraging cooperation among manufacturers, and on October 18, 1922, the British Broadcasting Company, Ltd., was established as a private corporation. Only bona fide manufacturers were permitted to hold shares, and the directors of the firm, all of whom represented manufacturing interests, met under an independent chairman. The company's revenue came from half of the 10-shilling license fee for receivers and a 10-percent royalty on the sale of receiving sets and equipment. Provincial stations were provided for, and all stations were to broadcast "news, information, concerts, lectures, educational matter, speeches, weather reports, theatrical entertainment. . . ."

Already several precedents had been established that were later followed in many other countries; of these the license revenue was the most important, but the royalty on sets and equipment was also adopted elsewhere, even after its abandonment in Britain. Because the British Broadcasting Company was a monopoly and because British radio as a result developed in a more orderly manner than elsewhere, such problems and issues of broadcasting as control of finance, broadcasting of controversy, relations with government, network organization, and public-service broadcasting became apparent, and solutions were sought in the United Kingdom earlier than elsewhere.

In 1927, upon recommendation of a parliamentary committee, the company was liquidated and replaced by a public corporation, the British Broadcasting Corporation (BBC), answerable ultimately to Parliament but with day-to-day control left to the judgment of the Board of Governors appointed on the basis of their standing and experience and not representing any sectional interests. A key figure, the chief executive of the original company and director general of the corporation, was John Reith (later Lord Reith), whose concept of public-service broadcasting prevailed in Britain and influenced broadcasting in many other countries. The BBC retained its monopoly until the creation of the Independent Television Authority (ITA) in 1954. The BBC experimented with local radio in the late 1960s and expanded the number of local stations in the early 1970s. In 1972 the ITA became the Independent Broadcasting Authority (IBA), which assumed responsibility for establishing and regulating independent radio and television stations. Regional and network production companies are appointed by the IBA; the companies sell advertising time, but advertisers are not allowed to sponsor programs.

Radio developments in other countries. Even before the pioneer station in Pittsburgh commenced operations, regular broadcasts began from The Hague, running from November 1919 until 1924. In Canada the first regular broadcasts from Montreal began in 1920, while in Australia a small station in Melbourne opened in 1921, though the official start occurred in Sydney in 1923. In New Zealand several low-powered stations were operating in 1921, though the Radio Broadcasting Company was not founded until 1927. In Denmark experimental amateur stations went on the air in 1921, and the official State Broadcasting System was instituted in 1925. France began regular transmissions from the Eiffel Tower in 1922, and the first Soviet station commenced broadcasts from Moscow in the same year. By the end of 1923 there also were radio stations established in Belgium, Czechoslovakia, Germany, and Spain. The list of countries lengthened rapidly, with Finland and Italy beginning broadcasts in 1924 and Norway, Poland, Mexico, and Japan in 1925. In India organized broadcasting began in 1926; the Indian Broadcasting Company had stations in Bombay and Calcutta in 1927.

The problem of control

In most of these countries, the problem of control arose. In some countries private enterprise was given free rein, subject to licensing by a government department or agency and to agreement upon the wavelengths or frequencies to be used. In others there was closer control (*e.g.,* France) or encouragement to cooperation between potentially conflicting interests (*e.g.,* Germany and Japan). Britain's example was followed in Denmark, Sweden, several Commonwealth countries, and some British colonies. In Canada and France state and private enterprise operated side by side. Private stations were well established in Canada, for example, before the Canadian Broadcasting Commission was formed in 1936.

In France the Administration of Posts and Telegraphs handled early broadcasts; although a state monopoly was declared in 1923 and state broadcasting remained a department of the Administration of Posts and Telegraphs until World War II, some private stations were granted licenses, including Radio Normandy, which broadcast to the United Kingdom. Some of these private commercial stations continued in operation, broadcasting under government control until 1945, when their licenses were withdrawn and radio became a complete state monopoly, independent of the Administration of Posts and Telegraphs but answerable to the government.

In Germany the Ministry of Posts controlled and owned all technical equipment, while private companies started programs in various cities. Soon the Reich Broadcasting Company acquired controlling interests in these companies; in 1932 all were nationalized.

International conferences. The wavelength problems that created so much confusion in the United States and provided a strong argument for monopoly in Britain also arose internationally, particularly in Europe, where the concentration of heavily populated and advanced sovereign nations compelled international agreement. Telegraphy had led to an early conference in Paris in 1865 that created what became later the International Telecommunications Union. This event was followed by the Berlin conference of 1885 to discuss international telephone communications, two further conferences in Berlin in 1903 and 1906 on radiotelegraph, and still another in London in 1912 to cover the whole field of radio communications. An informal conference of 10 countries held in London in 1925 created the Union Internationale de Radiophonie. The union was based in Geneva, with a BBC representative as president and another as secretary general, and was the first international broadcasting organization. The use of wavelengths, copyright problems, and international program exchanges inevitably were discussed, and a plan was drawn up.

Agreement on wavelength allocation, implemented in November 1926, was based on a formula involving area, population, and the extent of telephone and telegraph traffic. In spite of its dominating position, the BBC, which had been using 20 medium wavelengths, emerged with one long wavelength, 10 medium wavelengths, and five further medium wavelengths shared with others but below the Post Office limit range for broadcasting of between one megahertz and 600 kilohertz (300 and 500 metres). (Long waves range from 30 to 300 kilohertz; medium waves from 300 kilohertz to three megahertz; and shortwaves from three to 30 megahertz.) All of the more advanced of the participating countries (which had risen to 16: Austria, Belgium, Czechoslovakia, Denmark, Finland, France, Germany, Hungary, Ireland, Italy, The Netherlands, Norway, Spain, Sweden, Switzerland, and the United Kingdom) had to make some sacrifices, and some, such as the United Kingdom, had to persuade their post offices to agree to the use of wavelengths outside the broadcasting range; but the principle of international agreement had been established. The Washington Conference of 1927 widened the area of cooperation in respect to radiotelegraph, broadcasting, and the international allocation of wavelengths, or frequencies. It was followed by the Madrid Conference of 1932, which codified the rules and established the official international frequency list. This agreement stabilized the situation until World War II, after which the European scene was substantially changed, and a conference in Copenhagen in 1948 reallocated frequencies in the European Broadcasting Area. The Atlantic City Conference in 1947 had already created the International Frequency Registration Board. A conference in Buenos Aires in 1952 prepared the text of the International Telecommunications Convention. The text was revised at Geneva in 1959, where radio regulations were also revised. Geneva also was the site of the 1963 conference for the allocation of frequency bands for space and Earth–space communications.

International organizations. The International Telecommunications Union, created in 1865, has worldwide membership. In 1947 it became a specialized agency of the United Nations. Within the union are the International Frequency Registration Board, the International Telegraph and Telephone Consultative Committee, and the International Radio Consultative Committee. Apart from the International Telecommunications Union, a number of organizations have been established, primarily on a regional basis, since World War II. When tensions between the East and West made the Union Internationale de Radiophonie almost unworkable, a strong organization, the European Broadcasting Union, was created by the countries of western Europe in 1950, and it has a membership that now includes not only all nations of western Europe but also Cyprus, Greece, Iceland, Israel, Lebanon, Morocco, Tunisia, Turkey, and Yugoslavia; associate members include the United States and most Commonwealth and former French colonial countries, as well as Japan and several Latin-American countries. A parallel organization, the International Radio and Television Organization, was created in 1950 to serve nearly all Communist countries (excluding Yugoslavia) and allies of the Communist bloc. The European Broadcasting Union is based in Geneva, while the International Radio and Television Organization has its headquarters in Prague.

Asian and African organizations

The Asia-Pacific Broadcasting Union, which was formally established in 1964 as a union of national broadcasting organizations in Asia and the Pacific, includes Japan, Australia, New Zealand, and the Philippines, as well as Iran, Turkey, Egypt, and most of the non-Communist countries of Asia; its headquarters are in Kuala Lumpur, Malaysia. The Union of National Radio and Television Organizations of Africa, which was formed in 1962, includes most former French and British colonies. The union is based in Dakar, Senegal, and has its technical centre at Bamako, Mali. The Arab States Broadcasting Union was formed in 1969 as an intergovernmental organization within the framework of the Arab League; the secretariat is in Cairo, and the technical centre is located in Khartoum, The Sudan. The Asociación Internacional de Radiodifusión primarily covers North, Central, and South America but includes some European countries. Its central office is in Montevideo, Uruguay. The Commonwealth Broadcasting Association, established in 1945 as a standing association of national public-service broadcasting organizations in the independent countries of the Commonwealth, bases its secretariat in London. The North American National Broadcasters Association, with its headquarters in Ottawa, Canada, began as an ad hoc group in 1972 and became a formal organization in 1978. Its members are Canada, Mexico, and the United States. The Caribbean Broadcasting Union is headquartered in Christ Church, Barbados, and is an association of National Broadcasting Systems of the Commonwealth, Caribbean, and Other Regional States. The International Broadcasting Society was formed in 1985 to improve the information flow between Third World and advanced countries and to foster cooperation among developing countries. Its headquarters are in Seoul, South Korea.

There are other international broadcasting bodies, including the United Nations Department of Public Information and the Culture and Communication Sector of the United Nations Educational, Scientific, and Cultural Organization (UNESCO). The International Broadcast Institute, created in 1968 as a nonprofit and nongovernmental association supported by charitable foundations, with headquarters in London, fosters a free flow of communications for informational, cultural, and educational purposes. There are also a substantial number of religious broadcasting bodies, some of regional and some of worldwide proportions; among the most important are the World Association for Christian Communications, set up in 1968 and based in London, and the Association Catholique Internationale pour la Radio, la Télévision, et l'Audiovisuel, based in Brussels. Radio Free Europe, based in Munich and financed by U.S. government funds, was established to broadcast pro-Western propaganda to the Communist countries of eastern Europe.

TELEVISION BROADCASTING

First practical television system

Early developments. Through a series of technical developments in Great Britain, Europe, the Soviet Union, and the United States, television reached a state of technical feasibility by 1931. In that year a research group was established in Britain under Isaac (later Sir Isaac) Shoenberg, an inventor with vast experience in radio transmission in the Soviet Union. He fostered the evolution of a complete and practical television-broadcast system based on a camera tube known as the Emitron and an improved cathode-ray tube for the receiver. Shoenberg saw the need to establish a system that would endure for many years, since any subsequent changes in basic standards could give rise to severe technical and economic problems. He therefore proposed a system that, though ambitious for its day, was fully justified by subsequent events. Shoenberg's electronic scanning proved far superior to the mechanical scanning method that had been developed by the pioneer John Baird. The government authorized the BBC to adopt Shoenberg's standards (405 lines) for the world's first high-definition service, which was launched in London in 1936. So adequate were they that they formed the sole basis of the British service until 1962, when they gradually were superseded by the European continental standard of 625 lines. The first notable outside broadcast by the BBC was the procession of the coronation of King George VI from Hyde Park Corner in November 1937; a portable transmitter mounted on a special vehicle made its first public appearance. Several thousand viewers saw the transmission.

Television developments were slower in the United States. It was not until April 30, 1939, at the opening of the New York World's Fair, that a public demonstration was made by the National Broadcasting Company, which announced in that year that it was ready to begin broadcasting for two hours per week. The Columbia Broadcasting System and the Dumont network began telecasting in 1939 and 1940, respectively. By mid-1940 there were 23 television stations in the United States. World War II, however, brought nearly all activity to an end, as electronics factories were converted to wartime production. The Federal Communications Commission had authorized only limited commercial operation (the first sponsored television broadcasts began in 1941), and gradually stations closed down; only six were left with limited programs to serve the owners of about 10,000 sets. When wartime restrictions governing the manufacture of receivers were removed in 1946, the stage was set in the United States for a rapid growth of the television-broadcasting industry. By 1949 there were 1,000,000 receivers in use; the 10,000,000 mark was passed in 1951 and the 50,000,000 mark eight years later. In England the BBC's television service was resumed in June 1946; by 1949 there were 126,567 television licenses, and by 1950 there were 343,882, equal in proportion to 1,000,000 in the United States. Other nations did not begin television broadcasting on anything resembling a wide scale until the 1950s.

Transoceanic broadcasts. The initial attempt to interconnect the television networks of Europe and North America came in 1962, when the American Telephone and Telegraph Company used its satellite, Telstar, to relay television signals between Andover, Maine; Goonhilly Downs, Cornwall; and Pleumeur-Bodou, Brittany. The first transmission, of a purely experimental nature, originated in the United States on July 10, 1962, and this was followed the next day by transmissions to the United States from France and England; the first colour transmission occurred on July 16. Reception was limited to about 15 minutes, the period during which the satellite was within sight of the sending and receiving stations. To maintain continuous transmissions, the planners of the system proposed using a series of satellites, so that at least one would always be in position to relay signals. In the mid-1960s, however, an alternative technique came to the fore: a single relay satellite in a "stationary" orbit, so adjusted that it would remain always above the same point on the surface of the Earth. The first public demonstration of this system was on October 10, 1964, when television coverage of the opening ceremonies of the Olympic Games was relayed from Tokyo to North America via a Syncom satellite positioned above the Pacific Ocean. The so-called synchronous communications satellite maintained an altitude of about 23,000 miles (37,000 kilometres), its position fixed with respect to the Earth, their periods of rotation being identical. In the early 1970s, such satellites were so placed that virtually any area of the Earth was within reach of any other by space-relay circuits. The transmitters and receivers used in space are capable of carrying many television channels simultaneously, in addition to telephone and other communications. The landing on the Moon by the American astronauts in 1969 was carried by satellite to an estimated audience of more than 100,000,000 viewers (see also TELECOMMUNICATIONS SYSTEMS).

Broadcasting the Olympic Games

Broadcasting systems

THE BROADCASTER AND THE GOVERNMENT

Most observers recognize that no broadcast organization can be wholly independent of government, for all of them must be licensed in accordance with international agreements. Although broadcasters in democratic countries pride themselves on their freedom in respect to their governments, they are not always free of stockholder or advertiser pressure, nor are producers and editors truly independent if senior executives, under pressure from whatever source, interfere with their editorial functions. Independence, therefore, is a relative term when it is applied to broadcasting.

In a monograph that was written for the European Broadcasting Union, broadcasting systems are classified under four headings: state operated, those that work under the establishment of a public corporation or authority, those whose systems are a partnership blend of public authorities and private interests, and those under private management. A brief summary of these systems provides an indication of the complex variations that have arisen.

State operation. Grouped under this heading are broadcasting systems that are operated by a government department or delegated to an administration, perhaps with a legal personality and even possibly independent in financial and administrative matters, but subject to the government and not essentially autonomous. Under this heading is listed the Soviet Union, where a special committee was set up in 1957 to be in charge of Soviet radio and television under the direct authority of the U.S.S.R. Council of Ministers. Similar arrangements were made in Czechoslovakia and Poland, except that the committees were given a legal personality. Romania has delegated broadcasting to a committee attached to the Council of Ministers. All-India Radio is a department of the Ministry of Information and Broadcasting. Similar arrangements have been common in most Communist countries and in most countries that were colonies but have gained their independence since World War II.

Establishment of a public corporation or authority. The BBC has been the prototype of this kind of system. Provided it abides by the charter and terms of the license under which it operates, the BBC has maximum independence as regards the disposal of its funds (although its revenue is subject to governmental decision as to the cost of the license that is required for every television or radio receiver), the production and scheduling of programs, and, above all, editorial control. Certain residual government powers are either hedged around with agreed provisos or never exercised. Its income, save for profits on the sale of programs abroad and the sale of various phonograph records and publications, is exclusively derived from licenses. External broadcasting (*i.e.*, broadcasting to areas outside national boundaries) is separately financed. The chairman and Board of Governors constitute the legal personality of the BBC; they are chosen by the government, not as representatives of sectional interests but on the basis of their experience and standing. Political parties in office have been careful to avoid political prejudice in these appointments.

Canadian Broadcasting Corporation

The Canadian Broadcasting Corporation (CBC)—or Société Radio-Canada—also has substantial independent powers as determined by the Broadcasting Act of 1958, though over the years producers often have been restive due to political pressures, and Canadian broadcasting has been affected by the problems arising from the aspirations of some French-speaking Canadian citizens. The CBC is dependent on an annual parliamentary grant for its finance, supplemented by an income derived from advertising that amounts to nearly a quarter of its annual revenue. Canadian broadcasting as a whole is a mixed system, with private broadcasting companies operating alongside the CBC.

The Japan Broadcasting Corporation, like the BBC, once had a monopoly on broadcasting but has since lost it. By a series of acts in 1950 the Nippon Hōsō Kyōkai (NHK) was charged with the task of conducting "its broadcasting service for the public welfare in such a manner that its broadcasts may be received all over Japan." The NHK Board of Governors is appointed by the prime minister with the consent of both houses of the Diet. The system is financed almost exclusively from the sale of licenses for receiving sets.

Though West German broadcasting is properly included in this category, the situation there is substantially different, for the basic radio and television services are not a matter for the federal government but for the individual states (*Länder*). The nine state broadcasting organizations are also grouped together in a national organization to cooperate in the first television network. In each state, though there are some variations, there is a broadcasting council that is appointed by the legislature or nominated by churches, universities, associations of employers or trade unions, political parties, or the press; an administrative council; and a director general. Their revenue comes from receiving-set licenses and sometimes also from advertising.

Belgian broadcasting

The broadcasting system in Belgium provides an interesting example of a device that has been used successfully for coping with a two-language country. There are three public authorities: one for French broadcasts; a second for Flemish; and a third that owns the property, owns and operates the technical equipment, and is responsible for the symphony orchestra, record library, and central reference library.

Partnership of public authorities and private interests. In many cases this partnership is nominal and historical rather than substantial and actual. The outstanding example is Radiotelevisione Italiana (RAI), originally founded in 1924. In 1927 an agreement was made with the government for a 25-year broadcasting concession. The charter was extended to cover television in 1952, and two years later a government agency acquired control and more than 75 percent of the shares. A board of managers includes 10 members elected by a special parliamentary commission, and the remaining six are chosen by the General Assembly. Inevitably, therefore, RAI is subject to close government surveillance, but an ad hoc committee of 30, appointed from all political parties by the presidents of the two legislative chambers, is responsible for ensuring political independence and objective news reporting. The organization must also prepare an outline of programs on a quarterly basis for approval by the Ministry of Posts and Telecommunications, aided by an advisory committee concerned with cultural, artistic, and educational policies. A separate organization runs the broadcast advertising business, which, together with receiving-set licenses, provides the revenue of RAI. By the early 1980s there were about 500 private television stations operating in Italy alongside the RAI.

Ownership of Swedish radio

In Sweden the broadcasting monopoly is technically a privately owned corporation in which the state has no financial interest, thus emphasizing the independence of Sveriges Radio from the government. The shares of the corporation must be held by the Swedish press (20 percent), large noncommercial national bodies or movements (60 percent), and commerce and industry (20 percent). The board of governors is made up of a chairman and government nominees and an equal number elected by the shareholders; there also are two employee representatives of Sveriges Radio on the board. The government reserves the right to determine the amount of revenue from receiving-set licenses on an annual basis and thus controls both investment and the amount of broadcasting. The government, however, does not control how that revenue is spent. On balance, Sveriges Radio has a substantial measure of freedom.

In Switzerland, too, there are elements of partnership between private interests and public authorities, but the federal constitution, the need to broadcast in three languages, and geographical factors have led to a system by which the Swiss Broadcasting Corporation is composed of three regional societies based in Lausanne, Zürich, and Lugano-Besso.

Private management. Most of the broadcasting organizations under this heading are commercial firms that derive their revenue from advertising, which takes the form of brief announcements scheduled at regular intervals throughout the day. In some cases a program, such as a sports event or concert, may be sponsored by one advertiser or group of advertisers. Methods and degree of government control vary, and no general characteristics may be isolated. Private-enterprise radio predominates in the United States and Latin America.

Subject to similar controls in these countries are many nonprofit educational stations, financed by universities, private subscriptions, and foundations. There is a public-service network, the Public Broadcasting Service, in the United States.

Other methods of distributing sound and vision programs by wire and cable are not strictly broadcasting. In the main, wire-diffusion enterprises concentrate on giving efficient reception in densely populated areas, large blocks of buildings, and in hotels of programs that are being broadcast. A tall apartment building, for example, may have one television antenna on its roof to which residents may attach their receivers. "Pay TV," the provision of such programs as sports, special events, films, and theatrical performances via direct cable lines to subscribers, has been growing in popularity. Cable television reached about 21,000,000 homes in the United States by the early 1980s and has created two industries in broadcasting: one to hook up homes, the other to supply the programs. Cable television has drawn viewers away from the major commercial television networks, whose share of the prime-time audience has fallen and is expected to decline further.

THE BROADCASTER AND THE PUBLIC

Nature of the broadcast audience. The psychology and behaviour of a radio or television audience, which is composed principally of individuals in the privacy of their own homes, differs considerably from that of an audience in a theatre or lecture hall. There is none of the crowd atmosphere that prevails in a public assembly, and listeners are only casually aware that they are actually part of a large audience. This engenders a sense of intimacy that causes the listener to feel a close personal association with the speaker or performer. Furthermore, many people will not accept in their own homes many of the candid forms of expression that they readily condone or support on the stage or in literature.

Because it owes its license to operate to the state, if indeed it is not state operated, and because of its intimate relationship to its audience, broadcasting functions in a quasi-public domain, open in all its phases to public scrutiny. It is therefore held to be invested with a moral as well as a legal responsibility to serve the public interest and must remain more sensitive to public sentiment and political opinion than most other forms of public expression.

Audience measurement. For economic reasons, as well as those outlined above, evaluation of audience opinion and response to radio or television programs is important to the broadcaster. Audience measurement presents difficult problems, because there is no box office by which to determine the exact number of listeners. Mail received comes principally from those who have the time and inclination to write and cannot be regarded as wholly representative. Audience-measurement information may also be obtained by telephone-sampling methods, interviews in the home by market-research organizations, or by special recording devices attached to individual receiving sets. The latter, installed with the owner's consent, record the amount of time the set is used, when it is turned on and off, and the stations tuned in. These devices are expensive, however, and do not necessarily indicate whether someone is actually watching or listening, and they are therefore limited to small samples of the total audience. Whatever the method of rating, commercial broadcasters are quick to alter or discontinue any program that shows lack of audience appeal, and the listeners are thus influential in determining the nature of the programs that are offered to them. In commercial broadcasting, sponsored programs also are affected by their apparent success or failure in selling the goods advertised.

Audience-measuring devices

Educational broadcasting. It is difficult to give an account of educational broadcasting in countries where broadcasting is largely or wholly a matter of private management and where the larger and more important stations and networks are private commercial enterprises. Nevertheless, considerable numbers of educational transmissions are made in the United States and Latin America by universities and colleges and sometimes by municipal or state-owned stations. The Public Broadcasting Service in the United States has increased the amount of educational and generally more thought-provoking material available on the air, and in Latin America some countries use broadcasts not only to support the work of teachers in schools but also to combat illiteracy and to impart advice to isolated rural populations in matters of public health, agricultural methods, and other social and practical subjects. The Roman Catholic Church has been in the forefront of the latter activity, operating, for example, the Rede Nacional de Emissôras Católicas in Brazil and the Acción Cultural Popular in Colombia. A similar use of broadcasting is made in most of the tropical countries of Africa and Asia.

Japan's educational broadcasting

Japan's NHK has the most ambitious educational-broadcasting output in the world. Each of its two television and AM radio services is devoted wholly to education, while general television services and FM radio also transmit material of this nature. Japan prepares programs for primary, secondary, and higher education, special offerings for the mentally and physically handicapped, and a wide range of transmissions under the general heading of "social education," which includes foreign languages, vocational and technical instruction, advice on agriculture, forestry, fisheries, and business management, plus special programs for children, adolescents, and women. The educational broadcasts of NHK reach more than 90 percent of Japan's primary and secondary schools.

In Europe the French state broadcasting service devotes more than one-half of its radio output to educational and cultural broadcasts in the arts, letters, and sciences; and on television about 14 percent of its first and second networks are devoted to adult education. Primary and secondary instruction is offered, as are refresher courses for teachers and university-level courses.

Although Italian radio devotes less than 1 percent of its output specifically to educational programs for children, nearly 20 percent is given to cultural and allied offerings. Educational television began in Italy in 1958 with courses of a vocational nature, followed by transmissions aimed at secondary schools. In 1966 special programs were initiated for areas where there are no secondary schools. By the early 1980s, 17 percent of Italian television time was devoted to educational and school broadcasts and 4 percent to cultural programs.

Swedish radio offers a comprehensive service of educational and cultural broadcasting, with the output on television higher than that on radio. There is also a substantial output of adult education at the primary, secondary, and university levels, with about 1,400 school broadcasts a year, and Sweden has concentrated on vocational training and refreshment for teachers. German broadcasting, by contrast, has been used much less for formal education. In The Netherlands more than two and a half hours of school and continuing education broadcasting are broadcast weekly on the radio; in addition, nearly eight hours of educational television are transmitted every week.

The BBC pioneered in education; its work, in both radio and television, has steadily expanded. The BBC offers primary and secondary students more than 100 radio series and nearly 40 television series. The BBC also offers a wide range of biweekly programs especially designed for study in degree courses with the Open University, created and financed by the government, with the broadcast teaching supplemented by publications and correspondence work. By the mid-1970s, BBC broadcasts for the Open University averaged 16 hours weekly on radio and more than 18 hours on television. In addition, the Independent Broadcasting Authority in the United Kingdom has required the commercial-program companies to contribute educational material both for schools and for adults; by 1970 this amounted to 10 hours weekly during periods totalling 28 weeks of the year.

In Australia there is a small educational output on the commercial stations, both radio and television, but by far the greater part of educational broadcasting is undertaken by the Australian Broadcasting Corporation. Educational programming accounts for about 4 percent of radio time and 18 percent of television output, the majority of which is broadcast to schools and kindergartens. The Canadian Broadcasting Corporation is required to provide educational programs in both English and French and does so on its AM and FM radio networks, as well as on television.

Broadcasts for external reception. International broadcasting—the transmission of programs by a country expressly for audiences beyond its own frontiers—dates from the earliest days of broadcasting. The Soviet Union began foreign-language transmissions for propaganda purposes in the 1920s. Fascist Italy and Nazi Germany made such broadcasts at a later date. France, Great Britain, and The Netherlands were next in the field among European countries, though their first use of shortwave broadcasting was aimed at French-, English-, or Dutch-speaking populations overseas. Great Britain began foreign-language broadcasting early in 1938 with a program in Arabic and transmissions in Spanish and Portuguese directed to Latin America. By August 1939, countries broadcasting in foreign languages included Albania, Bulgaria, China, France, Germany, Great Britain, Hungary, Italy, Japan, Romania, the Soviet Union, Spain, the United States, and Vatican City.

Radio propaganda

During World War II foreign-language broadcasting continued; the programs of the BBC in particular, because of their reliability and credibility, had an important effect in maintaining morale among the countries that were under German occupation. The continuance of international tension after World War II led to remarkable growth of foreign-language services. In 1950, for example, all of the Communist countries of eastern Europe except East Germany had launched external services, although these were on a small scale, and even the Soviet Union was transmitting a total of more than 500 hours of broadcasts weekly in all foreign languages. The United Kingdom's output, which had once led the field, had been reduced to slightly more than 600 hours a week and the Voice of America to less than 500 hours per week. By the early 1980s the situation had changed radically. The Soviet Union alone broadcast more than 2,000 hours per week, and the output of all Communist countries of eastern Europe (excluding Yugoslavia) totalled about 1,500 hours. The United Kingdom logged 744 hours in 1981, West Germany 785 hours, and the United States broadcast over the Voice of America and Radio Free Europe/Radio Liberty 1,925 hours a week. The output of China had risen from 66 hours weekly in 1950 to 1,375 hours by 1981. The increase in Chinese broadcasts reflected in part the rising tension between China and the Soviet Union; significantly, the output of China's ally for much of this period, Albania, rose from 26 to 560 hours weekly during the same period. By the early 1980s Japan was transmitting for 263 hours, while Australia and Canada also sponsored external broadcasts.

Monitoring and transcriptions. A logical development following from external broadcasting is the monitoring of foreign broadcasts and their analysis for intelligence purposes. The BBC in particular has a highly developed monitoring service; this activity often yields valuable information. The Central Intelligence Agency of the United States is also active in monitoring and analyzing foreign broadcasts. Transcriptions (recordings) of programs produced in either the domestic or the external services of one country can be acceptable for broadcast in others. Radiobroadcasts of an educational nature can be used in different countries speaking the same language. Although many radio transcriptions are supplied free, in television the situation is different, and there is a substantial trade in television films.

Pirate and offshore stations. In some countries where broadcasting in general or radio alone is a monopoly, radio has had to compete for brief periods with independent commercial stations mounted on ships anchored at sea outside territorial waters. Sweden, Denmark, The Netherlands, and the United Kingdom have been the countries most affected by these stations, which have made use of unauthorized wavelengths, thus endangering other radio communications and operating free of any copyright obligations in respect to any of their broadcast material. Government action gradually has forced closure of such operations: in Sweden a competitive service of popular music proved effective; and in Denmark naval police action (the international legality of which may be questioned), followed by confiscation and heavy penalties, brought an end to the pirate station. The United Kingdom combined legislation penalizing any party who advertised or supplied such ships with the launching by the BBC of Radio 1, substantially a popular music service, to solve the problem. The French have had a particular problem of competition from the so-called *postes périphériques,* which include Europe No. 1 in the Saar and Radio Andorra in the Pyrenees, not to mention the French-language broadcasts of Monaco, Belgium, Luxembourg, and Switzerland. The strongest competition came from Europe No. 1, in which the French government finally purchased a controlling interest. (J.A.Ca./Ed.)

The problem of *postes périphériques* in France

Broadcasting as a medium of art

The artistic potential of any medium is determined by the unique form it offers and forces on the artist and its capacity as an effective vehicle of communication in its own right. The form of any art includes the circumstance through which it reaches its public. William Shakespeare's stage was little more than an open platform on which any action he cared to represent could be compassed provided he gave his actors the necessary words to indicate the place, circumstance, and atmosphere of the action. But his plays would have been null as practical drama without the circumscribing enclosure of the Elizabethan circular theatre auditorium—the "wooden O"—which gathered the audience around the platform, sealing them off from the outside world and concentrating their attention on the performance. As active auditors they became an integral part of the drama, and one must be constantly aware of them in the very writing, structure, and timing of the plays. Shakespeare's art was born of the discovery of the potentialities of the actor–audience relationship.

THE ART OF RADIO

In a similar way the art of radio began to be discovered when those engaged in broadcasting became aware of the nature of the medium in which they were working and of their special relationship to their audience. The discovery took time. (The artistic potential of radio was not explored until the 1930s.) Radio was the only medium in which performers were invisible to their audience. Broadcasters tended at first to adopt the manner of the stage or the pulpit: thinking in terms of a mass audience, the inexperienced broadcaster gave his voice and style an artificial inflation totally unsuited to the new medium. His actual audience was composed of small groups and individuals, usually at home or in informal circumstances, often doing other things at the same time. The basic art of radio consisted in adapting manner and style to these new circumstances. Few programs could expect to take their audiences for granted.

News broadcasts were among the exceptions, and news broadcasts in most countries came to be delivered in a fixed, impersonal manner, the newsreader suppressing his personality as far as possible and adopting a "team" voice. Most other spoken radio formats required using the voice in such a way as to hold the attention of the listener, and this in turn meant recognizing the nature of the microphone as a medium. The public's span of attention was found to be limited. The news summary was allotted five or 10 minutes, while a talk might last 15 minutes, or in special instances up to 30 minutes. Much consideration was also given to the appropriate styles for various audiences, depending on the classes of listeners to whom broadcasts were addressed.

The art of radio emerged in Britain—and in certain other countries adopting the same "public service" approach—as a medium nominally addressed to everyone but actually resembling a kind of broad-based national journal with special sections addressed to specific interests and tastes, some more demanding on the intelligence than others. The popular radio talk (a form of spoken journalism, or essay, often excellently composed and delivered) was shorter and more informal in style than that of the "serious" or purely educational talk. Broadcasting offered unique opportunities for bringing the nation's highest intelligences into the living rooms of so-called minority audiences (often amounting to millions) who were prepared to listen to concentrated exposition and argument. From this developed, particularly in Europe, channels specializing for part or all of the day in minority interests. The listeners supported the service by paying an annual license fee. In the United States, on the other hand, privately owned broadcasting companies got their revenues from advertising and tied their programming to the advertiser's desire to reach the widest possible public. In Japan there were both public and commercial broadcasting services, the former being financed, as in Europe, by license fees from owners of receiving sets. In the Soviet Union broadcasting was recognized, in the words of *Pravda,* as "one of the most powerful weapons of the cultural revolution." Under Stalin virtually all receivers were wired to local exchanges so that the listener could choose only among approved programs. The service had to be regionalized since the U.S.S.R. includes populations speaking about 80 principal languages. Aside from news and commentary, the broadcasts were generally cultural rather than directly propagandistic.

The development of radio as an art form was thus dependent on the way it was organized and financed. There were rich new fields to be opened up in drama, light entertainment, and documentary programming, conceived specifically for the medium, while at the same time some traditional art forms (notably stage drama and music) were transmitted with success. Fiction and poetry reading also became a staple part of sound broadcasting.

THE ART OF TELEVISION

Television compared with film

In the 1950s and 1960s radio was overtaken by television. At first television was considered to be little different as a medium from the film. But although television was a hungry user of film, it needed film in forms that differed from those required by the theatres.

The difference between film and television as art forms stemmed from the physical and financial conditions governing production, distribution, and exhibition. The relationships between the media and their publics were also different. The initial difference lies in the cameras and their function in production. The film camera supplies a record on celluloid in the form of a two-dimensional image, which, suitably edited, can be subsequently projected onto a screen. The television camera accepts and makes available for immediate transmission a two-dimensional image that remains unrecorded and passes with the event, like the image in a mirror (though this image can, by using additional equipment, be recorded on film or videotape). The film camera is associated with a lengthy effort of photographing, cutting, editing, and dubbing—an elaborate process of selection and assembly that may involve months of work. Although television images may also be stored and edited through videotape, the essential television form is the immediate transmission to the public of events occurring at the moment—political and social events, news summaries, commentary, and discussion.

The basic art of television is the control of this immediate flow of images. They can be preselected insofar as the cameras may be set up at chosen vantage points; after that, however, the director must select among the images they give him. The director-editor uses his skill to secure an immediately effective flow of images from the multiple viewpoints his cameras and their lenses collectively represent. In the film the same end is achieved by the quite different process of fragmenting and recording the action piecemeal, thus creating a succession of images that can be subsequently put together by editing and dubbing.

Those who first struggled with the practical aesthetics of television attempted to see the medium on the one hand as a kind of visual radio and on the other as a form of "diluted cinema," a rather poor cousin of the theatrical film. This was in part because they came either from radio or from filmmaking and saw the medium in relation to their previous occupations. Writers, directors, and performers from radio tended at first to reduce the television image to a "talking head," with the addition of occasional still pictures, film clips, or cut-ins from other broadcasting stations. This was especially the case in countries in which television initially lacked adequate financing (such as France) and directors could not afford costly pictorialization. On the other hand, personnel coming from filmmaking were appalled at the speed with which they were required to prepare and mount their television programs.

Television differs most from film in its relationship to the audience. The film is an event designed for a theatre with an audience specially assembled for the performance. Television, on the other hand, resembles a private performance in the home. The attitude of a person sitting perhaps alone and often for hours on end before a comparatively small picture screened in the familiar surroundings of his living room is quite different from that of a person who has gone out to share the special audience experience of a theatre. The tension is more slack; concentration is constantly threatened by irrelevant interruption. Whereas one is absorbed by a good film in a theatre to the exclusion of all else, one merely "watches" television. The television audience is preoccupied not so much with an individual item as with the free flow of item after item. Television is like a talking picture magazine, going on daily and nightly, asking little, giving out along with its entertainment a quantity of easily assimilated information ranging from formal news coverage to informal, gossipy discussions of the lighter affairs of the day.

Television also differs from film with respect to its visual impact. In the movie theatre a highly magnified image fills the central part of the field of vision in an otherwise darkened hall, exciting curiosity and response to a degree far beyond that obtained by a standard-size television screen in a relatively undarkened, and much smaller, living room. The great, fully loaded images of the big screen, from D.W. Griffith's *Intolerance* of 1916 to the Russian *War and Peace* of 1968, have involved the investment of large sums of money in what is called "production value"—the accumulated content of those images with their crowds of people and their elaborate sets. Skilled viewers in the movie theatre perceive and appreciate an astonishing amount of detail. In comparing television, however, one has only to watch a film produced for big-screen theatre to realize the limitations of the small television screen, in which the actors, speakers, or commentators must occupy most of the visual field.

TECHNIQUES AND BORROWINGS

The various media compared

It is useful to view all of the media together, ranging from the individual performer appearing in the flesh before his audience to the complex presentations of the electronic and allied media. They may be compared in terms of the relationship of the performer to his audience as shown in Table 1. The media also vary in the kind of performance on which they can draw, either derivatively or creatively, as shown in Table 2.

Table 1: Relationship of a Performer to His Audience

performer	audience	relationship
individual speaker, storyteller, or singer	assembled group or audience in any place	direct; performer to audience
company of players, singers, dancers, or musicians	assembled audience in a theatre or concert hall	direct; in an enclosed area
performance by an actor or performer recorded on film; factual film with commentator	assembled audience in a motion-picture theatre, hall, classroom, or other formal place	remote; through a photographed and projected two-dimensional image on either a large or small screen, with recorded sound
radio broadcast by an actor, aural performer, newsreader, or commentator	dispersed audience, located mainly in their own homes	remote; through a signal broadcast in sound only
televised presentation by an actor, singer, dancer, performer, or commentator	dispersed audience, located mainly in their own homes	remote; through a two-dimensional image on a small screen, accompanied by sound

The tables make clear the extent to which the various media borrow from each other. Just as the Greek drama drew on ancient myths and legends and the Renaissance drama on classical and contemporary material alike, so the voracious demands of the new 20th-century media have driven producers and scriptwriters to acquire the rights to existent material in other media, particularly the novel and the drama. Radio and television have overlapped increasingly with journalism, many journalists becoming broadcasters and commentators.

But much of the borrowing has been mechanical and technical rather than artistic in nature. Radio broadcasting exploited the phonograph record as a means of preserving sound; in a similar way, television drew upon the film. The invention of magnetic tape for recording both sound and video signals has now linked together all of the mech-

Table 2: Relationship of the Medium to the Performance

medium	nature of presentation	type of material
rostrum	visible, audible performance by a single person	oratory, preaching, recital; the speech, sermon, song, reading, monologue, monodrama
live theatre; concert hall	visible, audible performance by a group or company	drama, opera, ballet, revue, circus, *etc.*, with or without music; the concert
motion pictures (silent)	visible, but not audible performance presented by means of cinematograph projection	mimed drama with titles, documentary presentation, news record, or animated film; presented with "live" sound (music, commentator)
motion pictures (sound)	visible, audible performance presented by means of a cinematograph projection	original screenplays or material adapted from theatrical, fictional, or other sources; according to degree of adaptation, the sound film supervenes on the form of its source, making something new; also news, factual, and documentary material
sound radio	audible but not visible broadcast performance	the whole range of human activity, from the news bulletin, report, commentary, discussion, talk, or actuality recording to the complete cycle of the audible arts—story and poetry reading, drama and documentary, music and opera, including material specially created for the medium
television	visible and audible broadcast performance	includes all of the above, but seen as well as heard

anized media—phonograph, telephone, radio, sound film, and television—and made available a virtually complete record of the sights, sounds, arts, and culture of modern society.

Preservation by recording is in itself not a creative art but a service to art created elsewhere. A principal function of radio and television broadcasting has been the dissemination of works of art created for other media. This is particularly true of radio; in television these works are more often transformed to meet the requirements of the medium and become different art forms. When an opera is performed in a television studio in a way that meets the potentialities of the electronic cameras, the result is television opera—a different form from stage opera. When an opera is commissioned and composed specifically for television (as was Benjamin Britten's *Owen Wingrave*), then television may be considered an artistic medium in its own right.

Storytelling

Dramatic techniques. Radio began by restoring the ancient art of the storyteller. Writers for radio next learned how to suggest place and time by word of mouth, accompanied by the impressionistic use of sound and music. Thus was born the genre of radio drama. The radio dramatist must address himself to the imagination of listeners who are unable to see what they are experiencing. This limitation carries with it a certain freedom. Just as Shakespeare's independence from stage decor left him free to move his action widely in time and space (*Antony and Cleopatra,* for example, has 42 wide-ranging changes of scene), so radio has been free to create its own plastic continuities of action and time-space reference. Radio has been highly creative in the fields of drama and documentary and also in quite new forms of imaginative light entertainment.

Television, on the other hand, adapted techniques already established by the sound film of the 1930s and 1940s. In the initial rivalry between film and television, economic and technical factors both played a part. The first television plays were like the simplest kind of film dialogues; they avoided elaborate sets or large casts, because the screen was not large enough and because they cost too much. Television material was highly expendable, like newspapers and journals that are discarded after a single use. Only gradually did the international distribution of selected television programs, particularly within the large Anglo-American market, permit more money to be spent on "production value" in television.

Television drama came into its own during the 1950s with the emergence of writers and directors who shook themselves free from the old models and began to develop their own techniques—an extension of the two-dimensional image with sound into fields that the cinema could not or would not enter. The creativity of television in the purely artistic sense lies in the unique opportunities it offers the maker. These opportunities were beyond the reach of the filmmaker, who had no way of impelling his sponsor to finance him in such ventures. Here art and the nature of sponsorship can be said to overlap, as is so often the case in the history of art.

Use of videotape

Film techniques. The basic principles that the television image shares with the film image are, of course, its freedom to select the compass of each individual shot and its freedom to determine the nature of the movement within it. The form of presentation depends in both cases on a continuity of such shots in order to build up a narrative flow. Film and television narrative are based on the same principles of mobile composition—the selective (or edited) flow of selective shots of the action. Despite the technological differences in their production, they are aesthetically closely linked and will continue to have a close relationship with each other. This relationship naturally extends into the technical field. Television adopted videotape in order to achieve an immediate high-quality record of the electronic image. This seemed at first to be a threat to the use of film in television, but that has not proved to be so; the film camera is indispensable in many branches of television production. On the other hand, filmmakers have found videotape to be useful in cinema production, since it provides the capability of checking the shot before the film is processed.

The development of television as an art form has not excluded its use as a channel for works produced in other media. On the contrary, production in other media increasingly has been financed out of revenues from its subsequent transmission on television. Since the earliest years of its existence, television has depended on the regular screening of a vast backlog of movie films. The high rentals paid on old films have induced television interests themselves to undertake the production of new films to be shown in theatres and subsequently on the television channels they operate. The feature films they produce often have relatively small casts and a higher ratio of in-close shooting, making them suitable for the smaller TV screen, just as most films now shot for wide screens keep the essential action in the centre so that they can later be shown on television. (R.M.)

Broadcasting operations

TYPES OF PROGRAMS AND DEVELOPMENT OF STUDIOS

There are a number of distinguishable types of programs that are broadcast, but they often overlap in technique, subject matter, and style. Radio, for example, broadcasts speech and music, but in an endless number of combinations. Television adds the visual element, greatly increasing the number of possible program forms. Most sizable broadcast organizations, however, have several categories for administrative convenience. But the definitions cannot be too precise, and lines of demarcation are necessarily vague.

Entertainment. Entertainment can include comedy, impossible wholly to differentiate from drama; quizzes, not always easily distinguished from relatively serious programs of information and education; popular music, in which the frontier with jazz and serious music is anything but rigid; and variety, or a series of unrelated acts, nearly always linked by a popular presenter or established performer.

From the early days of radio there was a tendency to make use of a variety format, and, as this approach represented an extension of old music-hall traditions, success was achieved by many programs in this vein. From the music-hall–variety-type program emerged the "gang show," in which a cast of performers remaining the same from week to week would make use of a series of humorous situations or catchphrases, gradually building up a familiar background against which the incongruities of the script could exploit humour to the full. A further development was the "situation comedy," in which a number of characters, such as the members of a family, remain in the same situation week after week but experience comic adventures. Though these laughter programs lost popularity on radio as television gained popular acceptance, they have become the mainstays of television. A contemporary phenomenon has been the comedy program involving substantial amounts of political and social satire. The situation comedy has also been influenced by this trend.

Situation comedy

The many types of comedy entertainment programs that are produced around the world all have one common characteristic: not only have the performers needed the stimulus of a studio audience, but also the listeners and viewers are stimulated by the laughter and applause of the audience. This has led to some abuses, such as the superimposition of laughter and applause on prerecorded programs, a practice that is frowned upon but still practiced. It has also meant that large studios are required to accommodate not only the performers, frequently including more than one music combination, but also the audience. In television there must be room for settings that have become increasingly ambitious and for dancers and choruses. Broadcasting organizations have generally been able to build studios of appropriate size, though radiobroadcasters in the early days preferred to purchase or rent small theatres.

In their form and structure, children's entertainment shows resemble those for adults. Animated cartoons, however, represent an exception to this rule; the Hungarians, Poles, and French have achieved genuine distinction in this area.

Drama. Radio and television drama is not best produced in a theatre; the nature of the studio is therefore different. Early radio drama was produced in a relatively small studio, often with a single microphone, just as early television plays were produced with a single camera. Radio engineers soon began to employ a control panel with inputs from more than one studio and sound effects ingeniously achieved; their counterparts in television expanded their use of cameras and sets. Mixing in radio from one studio to another and in television from one set to another, plus increasingly sophisticated sound effects and background music, have all become accepted techniques in drama production. Inevitably, television drama has borrowed substantially from the techniques of film production.

Feature films, usually originally made for the cinema, continue to form an accepted and important element in television schedules throughout the world. Both radio and television occasionally broadcast live stage plays from theatres, but there is a general feeling that such offerings do not adequately exploit the advantages of either medium. From the earliest days of radio and television, the studio-produced drama has been an important ingredient in program schedules; in television, as in films, it was not long before shooting on location also became an accepted practice. Offerings have included classical Greek drama, Shakespeare and other Elizabethan dramatists, the Spanish and French theatre, Russian and Scandinavian plays, and modern works.

Serial presentations on television and radio have included adaptations of famous works of literature, such as the novels of Charles Dickens, Honoré de Balzac, and Leo Tolstoy, the *Forsyte Saga* of John Galsworthy, historical costume dramas based on the lives of such figures as Henry VIII and Elizabeth I of England, and, of course, the romantic melodramas aimed largely at the daytime viewer or listener, known as "soap operas." Radio and television serials of fantasy and adventure are also produced for children.

Three other distinguishable types of drama have achieved almost universal popularity: western adventures; shows involving gangsters, crime, and police; and shows set in hospitals and other medical situations. Violent episodes in some crime and western adventure programs have drawn criticism from those who believe that such violence is harmful to children. In response, many broadcasting organizations have introduced codes of practice to minimize such scenes.

Violence on television

Western adventure programs, largely produced in the United States, have been popular with studios because of their relatively low production costs and ready salability abroad. Dramatic series of this type have been shown all over the world, often with dubbed sound tracks. Although these exported U.S. productions are often much less expensive than home-produced programs, Australia has been able to produce some western-type series, and Canada has exploited its legendary "Mounties." So many U.S. television programs have been exported, however, that broadcasting organizations in some nations, such as Japan and the United Kingdom, have taken steps to assure that home-produced dramas should have priority in terms of percentage of schedule hours and prime time (peak placing).

Spoken word. Spoken-word programs have included entertainment types, such as "This Is Your Life" and many of the "talk shows," in which a personality interviewer questions celebrities, sometimes with interludes of music or comedy or with serious discussions, documentaries, or lectures. A fear of controversy, the problem of maintaining an overall impartiality, and sometimes the belief that the mass audience would be alienated by programs demanding a conscious effort and concentration, combined, in the early days of radio, to limit the time given to serious spoken-word programs. It was not long, however, before many broadcasters developed a sense of pride and responsibility in their function and regarded it as their duty to provide information and opinion. In countries where broadcasting achieved a substantial measure of independence, some broadcasters gradually became concerned not only with the exposition of fact and controversy but also with the task of exposing the ills and abuses of their society.

News broadcasts

News continues to be the most important element in spoken-word radio. Since it was inescapable that broadcast news would affect the industry, newspaper proprietors in the early days of radio either made efforts to restrict the sources of news and the times at which it could be broadcast or sought themselves to enter the field. In areas where broadcasting was commercialized, the press was further concerned, because radio competed with it for advertising revenues and because radio could almost always get a story to the public before the newspapers could. Nevertheless, there is no evidence that radio news reduced the circulation of newspapers; some have even maintained that radio whetted the appetite of listeners for news and increased newspaper sales. It would seem, however, that television has adversely affected the daily press and, even more so, weekly or monthly magazines. Long before television outstripped radio, broadcasting organizations were employing reporters and special and foreign correspondents and were supplementing the service received from news agencies. Some broadcast reporters became public personalities in their own right.

Television news presented additional production problems; the announcer at the microphone reading from a script or TelePrompTer was not satisfactory, and it was not long before the greater part of television news was appropriately accompanied by relevant pictures. The need for film shots and the cost and difficulty of obtaining them were, and to some extent remain, serious problems. In spite of substantial expenditure on the supply of such shots, television news is open to the criticism that news values and objectivity are distorted by the availability or nonavailability of pictures.

Emotive effect of television news

In general, however, broadcasting organizations have adjusted to the much higher cost of television news. The syndication of film reports, the development of live networks on an international basis, such as Eurovision, and satellite communications have overcome most problems of news reporting on television. On the other hand, it has become apparent that the psychological impact of film shots of war and civil disturbance, as of accidents and disasters, is far greater than that of the radio report. Television reports of, for example, the Vietnam War did far more to influence public opinion than radio news bulletins could have done. Radio has the advantage, however, of not requiring the same degree of attention; the trend has been toward frequently repeated short bulletins. In the United States there are radio stations that restrict themselves entirely to news, usually in a continuous magazine format, plus, of course, the advertising spots. The news magazine, or newsreel, in radio was introduced even earlier on BBC. A series of brief reports, interviews, and extracts from speeches, making use of many voices and exploiting the technique of frequent renewal of stimulus, proved to be a successful formula. This technique has spread into news bulletins and is increasingly used in the coverage of current affairs, both in radio and television. In all these programs of news and comment, one of the problems has been that of the anchors, or presenters, and the degree to which they may be given freedom to project their personalities or express their views. In the United States there have been fewer inhibitions in this area than in countries where broadcasting is or has been a monopoly and where the need for and tradition of impartiality have been dominant. In the case of the BBC, newsreaders were long anonymous; but on television the identity of a newsreader, or of the presenter of a magazine of news or comment, cannot be concealed, and these inhibitions have broken down. Nevertheless, in western Europe and Commonwealth countries the impartiality of broadcasting services remains an issue of greater importance than in the United States or Latin America. In Britain, when the Independent Television Authority was created, it was enjoined to see that in the coverage of controversial matters each program should be balanced in itself. The BBC, with greater freedom, makes no effort to ensure balance in any one program, provided that an overall balance in respect of any issue is achieved over a reasonable period of time. In all developed countries elaborate programs are prepared to report the results of elections, though it is in the United States and the United Kingdom that these are most ambitious.

In radio straight talk persists in some countries, though less so than in the heyday of the medium. Nevertheless, some successful lectures at much greater length have been scheduled occasionally on television and in some countries on radio. Straight talk of 10 minutes or more does not lend itself to exciting television production, unless accompanied by filmed illustrations to the point where it all but becomes a documentary.

Another pattern popular in many countries involves a panel of distinguished figures under a chairman, answering questions of a topical nature from members of a studio audience. In some cases a parabolic microphone is employed so that questions may be asked from any part of the studio or hall in which the program is mounted; others may call for written queries in advance so that questioners can be conveniently seated in the first row. Some radio panel programs also solicit queries from members of the listening audience who call them in on the telephone.

The documentary program

Development of the radio documentary stemmed from drama, as writers searched for new material especially appropriate for broadcasting. Not surprisingly, early documentary was in dramatic form, and most of it was based on well-known historical events, of which the programs were in effect dramatic reconstructions. Production of radio documentaries was simplified by the invention of magnetic recording tape that was far easier to edit and use on location than its predecessors, the wax-coated disc and the wire recorder. Ironically, just when these technical advances had made the best form of radio documentary possible, the television documentary on contemporary themes began to supplant its radio counterpart. Documentaries have become more expository of public (current) affairs concerned with international relations, domestic politics, and social problems.

Religion. There have been, in the main, two types of religious program: devotional and information–discussion. The former comprises prayer, religious services, or hymn singing, either mounted in a studio or as outside broadcasts from a church, chapel, or hall. A third type is the dramatization of a religious theme, though the tendency has been to devote a good proportion of religious broadcasting time to documentaries, discussions, and interviews. Some sects have produced broadcasts that combine political and religious material. Missionary bodies, mostly under the control of one of the many international or regional religious broadcasting organizations, either buy time on commercial stations or operate stations in many parts of the world, including Latin America, Africa, and Asia.

Outside broadcasts. Although broadcasts do not constitute a distinct and definable form, they nevertheless have been since the birth of radio the most popular and arresting of all material transmitted on either medium. Sports of every description and ceremonial and political events have exercised an unfailing appeal and, in general, attract the largest audiences. Outside broadcasts have stimulated the imagination and taxed the ingenuity of television-broadcasting engineers to such an extent that they have accustomed the public to feats unimaginable to the pioneers of radio. The improvement of line communications, the development of mobile transmitters, and, above all, the use of satellite communications have given the outside broadcast an elasticity and an almost limitless range.

Effects on musical life

Music. Radio has had two important effects on the musical life of the world: it has widened the audience for all forms of music and has made easier the development of new forms, such as electronic music. Music remains a staple ingredient of radio in its own right, whereas in television, though there are programs of music as such, it is more often an adjunct to something else, as, for example, dancing, or as a small component of a mixed program. In the field of popular music, radio has immensely aided the rapid changes of fashion, which have coincided with technical advances in the making of recordings and their popularity and sales. A recognition by recording companies of the enormous power of radio in popularizing a song or performer has led to some abuses. In the United States "payola," or bribes, were given to prominent radio personalities by record companies in return for promotion of their songs.

The development of stereophonic sound techniques has revolutionized the record industry and has played an important role on radio, though earliest in the field of serious music. Frequency modulated (FM) radio broadcasts of serious music, and later of other forms of music, have been popular in many areas; some recordings are broadcast stereophonically but can be received on monophonic radios (see SOUND).

Opera, too, has profited from broadcasting, and outside broadcasts from opera houses, as well as studio performances on both radio and television, have done much in European and many other countries to bring this form of music to a large public. Music programs have presented more difficulties than most others in the matter of studios, partly because of the size of studio required for a full symphony orchestra and partly because of the delicate balancing of acoustics for proper reproduction of such performances.

RELATIONS WITH ARTISTS, SPEAKERS, AUTHORS, AND UNIONS

In the early days of radio, problems of fees, royalties, performing rights, copyright, and relations with unions rarely were regarded. Entertainers performed largely for publicity purposes. Only gradually did performers appreciate radio's effect, first, as a threat to their theatre earnings and, second, as a highly lucrative substitute. To try to trace how a modus vivendi was reached in these matters in different countries would present a picture of baffling complexity in the light of the different prevailing laws and different union structures. Generally speaking, copyright issues have revolved around the rights of record manufacturers and fees for composers. Rates and fees for reproducing recordings often have been the subject of disputes with the unions. Radiobroadcasters soon found that purchasing records or making their own recordings from live musical performances meant substantial economies; these, however, came at the expense of the musicians. Consequently, the musicians' unions sometimes attempted to prevent use of phonograph records or recording of live performances. In some countries, such as The Netherlands, the repeat problem has been solved by the performers receiving a fee for each repeat, the fee rising with each successive use of the recording until it ceases to represent an advantage to the broadcaster.

Copyright fees

Relations of the broadcasting organizations with their staffs have also been complex. In Canada attempts to exercise a restrictive control have led to revolts and resignations, while in France editors and producers who have been unwilling to conform to government policy have been removed from their jobs, though often under other pretexts. The position of staff is particularly vulnerable in those countries where broadcasting is a state monopoly: an example is Czechoslovakia, where senior broadcasting officials were ousted after the fall in 1968 of Premier Alexander Dubček, who had attempted to liberalize the Communist regime.

INTERNAL ORGANIZATION, ADMINISTRATION, AND POLICY CONTROL

Functions of the broadcast administration

The organization and administration of broadcasting bodies can, in the case of a small independent station, be relatively simple, and the policies can be implemented with ease. Sizable organizations, however, have a complex problem because it is not possible to determine success or failure purely on the basis of financial returns. Monopoly organizations, though in theory their sole purpose is public service, in practice often must take into account the views of the government. In the case of nonprofit–public-service operations dependent upon license fees for revenue but with commercial competition, ratings cannot be completely ignored, and these organizations must compete for mass audiences to some extent in order to justify their existence at the expense of the listening and viewing public.

The broadcasting administration has two essential functions: first, programming—*i.e.,* allocation of funds and setting of schedules—and, second, production, the preparation of programs. The former is in effect a branch of direction, and those in charge of planning program schedules and allocating funds have a power that if not checked can be absolute. On the other hand, these planners are dependent on the goodwill of the production and supply departments.

A main problem arises in the treatment of controversial subjects in the field of current affairs. Where broadcasters are under no obligation to be impartial, as in The Netherlands, or where, as in totalitarian countries, only one point of view may be aired, the problem does not arise. In democratic countries, however, where the broadcaster has independence and where there is a need to achieve an overall impartiality, the problem is very serious. Even though decisions may be reached by discussion and a consensus of opinion, the responsibility usually has to be carried by one person. No broadcasting organization has been able to find a complete solution to the problem that does not involve rigid control and intrusion on the independence of the editorial and production staff.

Administration must also deal with routine matters, such as staff pay and conditions of service, recruitment, finance, accounting, negotiations with unions, procurement of equipment, and provision of office and studio space. In general, it has been found best to subordinate such routine management operations to the needs of those directly concerned with the principal function of broadcasting. Much the same may be said of engineering and technical staff, though their research work and technical advances influence the decisions of direction and development of broadcasting.

THE STATE OF BROADCASTING IN SELECTED COUNTRIES

Argentina. Broadcasting in Argentina is wholly controlled by the government but only partly operated by government agencies. Partly for historical reasons, the method of control is not clear-cut: all broadcasting is subject to the approval of the Consejo Federal de Radio y Televisión (Federal Radio and Television Council), a body working under the State Secretary of Communications, though the Secretary of Information can and does intervene on behalf of the president of the republic. All political activities were suppressed between 1966 and 1971, and, even after the government of Alejandro Agustín Lanusse lifted the ban, the restored freedom was not reflected in broadcasting. Of the approximately 150 radio stations in the country, nearly half are grouped into two large networks: an official cultural noncommercial network with about 30 stations, some strategically placed in relation to broadcasts from other countries; and a government-controlled commercial network. There are more than 70 private commercial stations; some are small and low-powered. Television is primarily commercial, and many channels are administered by the state. LS82 Canal 7 in Buenos Aires is state-owned; the other three stations in that city have separately owned provincial affiliates. Argentina has more than 80 television stations, about 20 of them private.

Sources of Argentine programs

Early television in Argentina depended on U.S.-produced telefilms dubbed into Spanish, but today, though U.S.-dubbed feature films are still used, Argentine-produced programs dominate the market. Some material comes from Spain, Italy, and the United Kingdom. Argentine-produced programs can rarely be exported, because the Argentine accent in Spanish, particularly that of Buenos Aires, is not acceptable elsewhere in Latin America.

Australia. Australian broadcasting comprises four sectors: the national sector, the public sector, the commercial sector, and the Special Broadcasting Service. National broadcasting is the responsibility of the Australian Broadcasting Corporation (formerly the Australian Broadcasting Commission), which provides a wide range of programming—including educational, news, sports, religious, and entertainment—designed to promote Australian culture. Public broadcasting serves specific interest groups and is sometimes associated with a college or university. Its primary outlet is radio. Commercial broadcasting seeks a wide appeal, and almost 50 percent of it is locally produced, as required by law. The Special Broadcasting Service provides programming in more than 50 languages for Australia's ethnic communities. Both the Australian Broadcasting Corporation and the Special Broadcasting Service are unlicensed, publicly funded government instrumentalities empowered under the Broadcasting and Television Act of 1942. Public broadcasting is funded by subscription, sales of air time to community groups, and sales of publications. Public stations are not-for-profit. Commercial stations are funded primarily through advertising. Both public and commercial broadcasting stations are subject to licensing renewal review every three years. Until it was abolished in 1976, the Australian Broadcasting Control Board was responsible for all aspects of broadcasting in Australia. Since that time, however, the Australian government's Department of Transport and Communication has overseen that organization's planning and technical aspects, while an independent, statutory authority, the Australian Broadcasting Tribunal, is responsible for regulation and licensing as well as for determining programming and advertising standards.

The Australian Broadcasting Corporation operates about 240 radio stations and 360 television stations. In addition, Australia has about 140 commercial radio stations and 50 commercial television stations. Public broadcasting is heard on about 70 radio stations. The Special Broadcasting Service has two radio stations and two television stations and is Australia's only UHF (ultrahigh frequency) outlet. Radio Australia broadcasts in nine different languages to foreign countries, primarily in Asia and in the Pacific. It operates 13 shortwave stations. The Australian National Satellite System has been in operation since 1985 with the launching of AUSSAT-1 and AUSSAT-2. Through the Australian Broadcasting Corporation it provides television and radio broadcasting to homes in outback regions as part of the Homestead and Community Broadcasting Satellite Service. An additional satellite, AUSSAT-3, launched in 1987, supplements the program with a similar commercial service known as the Regional Commercial Television Service.

Brazil. There are more than 2,400 radio and about 180 television stations in Brazil; the majority are commercial. In general they are under the authority of the Ministry of Communications. All broadcasting is subject to censorship, and any station that runs counter to the government's wishes can be closed. There are radio stations of the Ministry of Education in Rio de Janeiro and Brasília; some of the states have official radio outlets, and a few have television installations. There are also some university radio stations and a few television stations, apart from the Roman Catholic educational radio network. The remaining stations are private commercial enterprises, operating independently or linked to one of the networks, of which the best known are associated with large newspaper concerns, such as Diarios Associados or O Globo. The larger radio networks use shortwave broadcasting, which permits simultaneous transmissions on medium-wave provincial stations. Provincial television stations prepare their newscasts, often in cooperation with a local newspaper and radio station. The others employ film, telefilm, and videotape to supplement local production. All radio stations must devote one hour each day to "The Voice of Brazil," a government news program supplied by the official Agencia Nacional; radio stations must also broadcast at least five hours a week of educational programs. Television stations may be called upon to broadcast programs produced by the Agencia Nacional, consisting mainly of government statements and ministerial and presidential speeches. In 1975 the government created Radiobras, the Brazilian Broadcasting Company, which broadcasts to the remote regions of the Amazon Basin, bringing those regions into closer contact with the political and cultural mainstream of Brazil. Television entertainment consists substantially of Brazilian-produced serials, supplemented by U.S.-produced and dubbed films and telecine programs. Brazil's first communications satellites, Brasilsat I and Brasilsat II, were launched in 1985.

Official news on Brazilian stations

Canada. Canadian broadcasting is overseen by the Canadian Radio-Television and Telecommunications Commis-

The Canadian Broadcasting Corporation

sion, which administers, regulates, and supervises the country's broadcasting. The principal broadcasting organization is the Canadian Broadcasting Corporation (CBC), which is financed primarily by public funds supplemented by television advertising. The CBC has two main television networks, one in French and another in English, and two main AM and FM radio networks, one in each language. In addition, there are small FM networks and some shortwave transmission. The CBC has more than 90 principal radio stations and more than 600 low-power relay transmitters; in addition, more than 60 privately owned stations are affiliated with it and are paid for transmitting its output. The CBC has more than 25 television stations and more than 600 rebroadcasters. There are more than 250 private affiliates and rebroadcasters. In addition to the CBC television network, there are four commercially operated networks: CTV Network broadcasts in English from coast to coast; TVA and the Réseau de Télévision Quatre Saisons broadcast in French across Quebec; and the Global Communications Ltd. network broadcasts in English in parts of Ontario. In 1972 Canada became the first country in the world to offer a domestic communications satellite system with the establishment of its satellite company, Telesat, and the launching of its first Anik satellites. Remote communities receive satellite broadcasts through its CANCOM program. External services are smaller than in most comparable countries; there are broadcasts on shortwave to the Canadian armed forces overseas and an international service, Radio Canada International, in English, French, German, Spanish, Portuguese, Czech, Slovak, Hungarian, Polish, Ukrainian, and Russian for listeners in Europe, Asia, Latin America, the Caribbean, the Middle East, Africa, and the United States.

France. Until it was disbanded in 1974, the Office de Radiodiffusion Télévision Française (ORTF), under the direction of the Minister of Information, oversaw all aspects of French broadcasting. It was replaced first by seven independent state-financed companies, then by a nine-member committee, followed in 1986 by the 13-member Commission Nationale de la Communication et des Libertés (CNCL). The CNCL was finally replaced by the Conseil Supérieur de l'Audiovisuel, which is charged with supervising all aspects of French broadcasting, including administration, programming, distribution of networks and frequencies, licensing, deregulation, and allocation of concessions for privatized channels. The following organizations are responsible for other aspects of French broadcasting: Télédiffusion de France operates the national transmitters; the Institut National de la Communication Audiovisuelle is responsible for broadcasting research and professional training as well as archival work; Radio Télévision Française d'Outre-Mer controls broadcasting to French overseas départements and territories; Société Française de Production et de Création Audiovisuelles is the national production company; Société France Media International is the national distribution company; and La Sept produces programs broadcast via satellite. Radio France comprises six national radio networks: (1) France-Inter, network A, a 24-hour service of entertainment and news, integrated with Inter-Variétés on regional transmitters and carrying programs produced by regional stations and France-Inter Paris, a morning program of popular music, and news flashes for Paris and the surrounding area; (2) Network B, a medium-wave network that includes Radio Bleue (programs for the elderly), programs for foreign workers, school programs, and university broadcasts; (3) France-Culture, network D, information and public affairs along with cultural programs; (4) France-Musique, network E, musical programming; (5) regional stations, network F, music and news, based in six major cities, along with Sorbonne R. France, with university lectures for the Paris region; and (6) France Info, 24-hour news and information. Radio France International broadcasts 24 hours daily to Europe, Africa, North America, Latin America, and Asia in seven different languages. There are also more than 1,500 private radio stations.

State-run television channels

Regular television service began in 1938, was interrupted by World War II, and recommenced in 1945. Colour television was inaugurated in 1967. France has two state-run television channels, Antenne 2 and France Régions 3, and four private stations. The formerly state-run Télévision Française 1 was privatized in 1987. Canal Plus was the country's first private channel in 1984. It is financed by subscription and broadcasts mostly films and sporting events. La Cinq broadcasts foreign light entertainment programs, sports, and films. M6 broadcasts general interest programs. TV5 broadcasts programs via satellite for Francophones in foreign countries. There are three private local stations. Cable television has been slow to develop in France and is primarily concentrated in the Paris region.

Italy. The origin and development of Radiotelevisione Italiana (RAI) is discussed above. Regular television broadcasts began in January 1954. RAI has three radio services on national networks on AM and FM: a First, or National, Program offering a balanced output; a Second Program essentially of entertainment; and a Third Program, which is educational. In addition, there is a substantial regional output. There are three television services: the first, on VHF (very high frequency), broadcasting 63 hours weekly, is the National Program; the Second Program also transmits 63 hours weekly on UHF; the Third Progam transmits 19 hours a week on Wednesdays and Sundays. The RAI provides limited regional television on special occasions only, except for a daily one and one-half hours in German for the German-speaking minority of the Trentino-Alto Adige region. RAI has 18 production centres and 21 regional offices. The Trieste and Bolzano offices are responsible, respectively, for radio output weekly in Slovene for the Friuli-Venezia Giulia region and, apart from the German television output, for broadcasts in German and in Ladin (Romansh) for the Trentino-Alto Adige region. RAI's revenue comes from a proportion, determined by the government, of the proceeds of the sale of radio- and television-receiving licenses and from advertising. Advertising is closely regulated and may not "prejudice the good quality of programs"; it is guided by a code, and the percentage of time given to it is limited. RAI devotes 70 percent of its radio output to light entertainment, 16 percent to news and information, 4 percent to cultural programs, and 1 percent to youth and educational programs. On television 36 percent is news and information, 19 percent is entertainment, 4 percent is cultural, and 17 percent is programs for schools and education.

Radio Roma broadcasts to Europe, Asia, Africa, the Middle East, the Mediterranean, Latin America, Australia, and New Zealand in 27 languages.

Japan. Nippon Hōsō Kyōkai (NHK) is the sole broadcast authority in Japan. A noncommercial, not-for-profit public station, it has two television and three radio networks (two AM and one FM), of which one television and one AM network are almost entirely devoted to education. The General Television program gives a balanced service, as does the First (AM) Radio service, while the FM (VHF) service is concerned mainly with cultural and local music programs. There is relatively little political broadcasting in Japan, and the first occasion on which candidates for election to the Japanese Diet were able to present their political views on television was in December 1969.

Japanese program scheduling

NHK has installed a system of computerized automation for its scheduling, resource allocation, and transmitting operations, probably the most advanced in the world, with a view to giving its production staff a maximum freedom for creative work. In 1978 the organization installed the 12,000,000,000-hertz (2.5-centimetre) wave for metropolitan area broadcasting and for broadcasting two television sound outputs simultaneously so that a single television image can be received with sound in either of two languages. NHK has 173 medium-wave transmitters for the First Radio network, 141 for the Second, and 474 for the VHF-FM network. General and educational television use about 2,800 transmitters each, most of which are rebroadcasting or relay stations. The so-called regional broadcasting is not so much regional as local and concerned mainly with news and practical information. Local television output averages one and a half hours, and local radio averages more than three and a half hours daily. In addition to NHK there are more than 130 broadcasting organizations that are members of the National Association of Commercial Broadcasters in Japan. They offer about 6,200 television stations and about 300 radio stations.

NHK is also responsible for Japan's external services, which are divided into a general service—*i.e.,* a worldwide service in English and Japanese that broadcasts daily—and a regional service for the Americas, Europe, Asia, Africa, and Australasia, in about 20 languages.

Mexico. There are about 850 radio stations in Mexico. Most are commercial, and more than 120 are grouped into two nationwide networks. There is a smaller number of television stations, constantly growing, many of which are grouped into networks of varying size. There are two nationwide networks, Televisa, which combined the former Telesistema Mexicano and Televisión Independiente de Mexico and is seen around the world, and the state-owned Imevision. Television, too, is nearly all commercial, though there are some university stations, of which the best known is Radio Universidad, run by the Universidad Nacional Autónoma de México in Mexico City; in addition, the Instituto Politécnico Nacional operates a cultural station in the capital. In theory there is no government control over broadcasting, but in practice there is no political broadcasting critical of the government or of the leading political party. Most stations carry the Hora Nacional, an hour-long officially produced program, every Sunday morning. The government has the right to use 12.5 percent of the total transmitting time of all radio and television stations. This right so far has not been fully exercised, though the announced intention is that the Comisión de Radiodifusión should acquire the resources to make full use of the time. Except for the few noncommercial stations, the majority of radio stations do little more than broadcast recorded popular music, news, and spot advertisements.

Television, with few exceptions, is substantially devoted to entertainment; a good proportion of material originates in the United States, though the government has banned some of the more violent shows. Production of programs for other Spanish-speaking countries is on the increase, with programs made for Puerto Rico, the Dominican Republic, Central American countries, Venezuela, Peru, Colombia, Ecuador, Argentina, Chile, and the Hispanic community in the United States. The government has proposed a network of officially operated small stations to provide cultural and educational programs.

The Netherlands. In all democratic countries, governments have found it difficult to reflect minority views in broadcasting. The Netherlands has made perhaps the most determined attempt to deal with this problem. The Dutch system basically consists of two national organizations, the Nederlandse Omroepprogramma Stichting, which is responsible for the transmission of general-interest programs and the promotion of Dutch broadcasting interests, and the Nederlandse Omroepproductie Bedrijf, an independent production company, and eight broadcasting societies (or organizations) that, through the size of their membership, have earned the right to produce a proportion of the foundation's output. The Broadcasting Act of 1966 called upon the responsible minister to allocate time on the air, in both radio and television, to bodies that fulfilled certain conditions, in particular a sufficient membership, and that, by October 1971, included: broadcasting associations or organizations, groups aspiring to recognition as broadcasting societies, churches, associations of a nature comparable to churches, political parties, other reputable associations of approved purpose, an advertising foundation, and educational bodies. As far as the full-fledged associations are concerned, the amount of time they have on the air is determined by their category, in turn dependent upon the number of subscribers, whose subscriptions pay for a weekly program bulletin. Some of the other bodies with time on the air may prepare their own programs or have them produced by groups with more experience. Organizations with at least 60,000 members may petition to broadcast from one hour per week to four hours. The government, however, has moved to restrict access to broadcasting, with legislation requiring aspirant broadcasting groups to offer innovative proposals to the existing range of programs in order to qualify. The financing of broadcasting, when production time is allocated among so many, presents a complex problem of accountancy. The revenue comes from the sale of receiving-set licenses and from advertising profits. Advertising on radio and television was first permitted in 1967 and is provided by the Stichting Ether Reclame.

Dutch broadcasting associations

The Netherlands has five radio networks and three independent television services. There is no regional television, but there are several regional radio organizations. The main categories of overall radio output are 25 percent news, public affairs, and information, 22 percent classical music, 14 percent light music, and 28 percent entertainment and other light programs. Television output is more diversified, with 32.3 percent entertainment (of which more than half is of foreign origin), 2.9 percent Dutch-produced drama, 5 percent films (mostly foreign), and 31.2 percent news, public affairs, and information.

Radio Nederland Wereldomroep (Radio Netherlands International) broadcasts daily shortwave transmissions to most areas of the world in Dutch and eight other languages (Portuguese, Spanish, Sranan Tongo, Papiamento, Arabic, English, French, and Indonesian).

New Zealand. In New Zealand the relatively small population means that broadcasting personnel are closely in touch with their audience, whose demand for a high-standard broadcasting service presents financial problems. The National Broadcasting Service, a government department set up in 1936, was faced by the competition of the National Commercial Broadcasting Service a year later. The two were amalgamated and reorganized as a government department, the New Zealand Broadcasting Service, in 1946. The service had some degree of independence from the start, and the inauguration of a television service in June 1960 provided the opportunity for the Broadcasting Act of 1961, by which was created the New Zealand Broadcasting Corporation, with a far greater constitutional autonomy. It is financed by license fees and advertising revenue. In 1977 the Broadcasting Corporation of New Zealand was created, incorporating two previously independent networks. Radio New Zealand has two radio medium-wave networks that include some broadcasts in Maori, Samoan, Tongan, Niuean, and Tokelauan. The corporation has more than 50 radio stations in more than 30 cities and towns. Television New Zealand operates two television networks, TV 1 and TV 2. It has more than 10 television stations with more than 600 relay stations, mainly low-powered. TV 3 is a private commercial station. The corporation is also responsible for the Foreign Service (Radio New Zealand), which broadcasts to Australia, the Pacific Islands, and the Ross Dependency in the Antarctic in English, twice a week to Samoa in Samoan, and once a fortnight to the Cook Islands, in Rarotongan, and to Niue, in Niuean. There are nine private radio stations.

New Zealand's external services

Soviet Union. The State Committee for Television and Radio Broadcasting operates a substantial undertaking under the Soviet Union's Council of Ministers. The chairman of the committee has four deputies, one each for television, external services, domestic radio, and for administration and finance, and there is an editorial board of 13 members. The committee controls output and is responsible for the equipment of television centres and for all personnel, but all lines, radio stations, and studios are under the control of the Ministry of Communications. Under the committee, domestic radio is run by seven Chief Editorial boards: program planning and presentation; propaganda; information (news); children; youth; literature and drama; and music, comedy, and satire. There is a joint radio and television department for sports. No program may go on the air without the approval of the editor in chief (or his deputy) of the appropriate editorial board. Though regional stations have a measure of autonomy, the committee controls the work of the regional committees handling radio in the various republics, regions, and districts. Radio in the Soviet Union requires between 300 and 400 transmitters for its networks and regional broadcasts. To achieve maximum coverage, it makes use of long waves, medium waves, and shortwaves, as well as FM. Moscow is responsible for five outputs as follows:

Program I is a mixed program covering the entire union.

Program II is a 24-hour musical service with news and commentaries every half-hour, also covering the entire population. Program III is primarily a musical and literary program for the European regions of the Soviet Union and reaches more remote regions by shortwaves and FM. There is also a local output of some three half-hours a day for the Moscow area on some of the channels used for Program III. Program IV is on FM and offers classical music nine hours a day. Program V is directed toward Soviets abroad. The regional effort is impressive, with broadcasts in about 70 languages in use within the territory of the Soviet Union. There are 23 principal regional stations. With so diverse an output there is no means of making a meaningful percentage breakdown of program categories, but Soviet radio on the whole devotes more time to information, educational, and cultural programs and less to entertainment than other countries.

The Moscow television station began broadcasting in 1939 and claims to have been the first European television station to renew operation after the World War II interruption. Colour was introduced in 1967. About four-fifths of the population are within reach of a television signal, for which there are about 900 main and 4,000 relay transmitters. In most of the principal cities there are at least two outputs to choose from. The satellites Molniya I and Ekran, combined with 90 Orbita ground and relay stations, greatly increased the size of the potential audience. At the Moscow television centre, Ostankino, there is a 1,739-foot (530-metre) tower used for television and other communications. The services broadcast cannot be easily analyzed, and it is claimed that from Moscow six can be broadcast simultaneously. There are four main channels: the First Program, all-Union since 1962; the Second Program, all-Union since 1982; the Third Program, educational at primary, secondary, and university levels; and the Fourth Program, which features locally produced programs in the various regional languages.

Soviet foreign-language transmissions

The external services described above are probably the largest of any country. The output can be divided into six types: (1) for foreign countries from Moscow, (2) for foreign countries from regional stations, (3) relays of domestic services for listeners abroad, (4) broadcasts for Soviets abroad, under the aegis of the Committee for Cultural Relations, (5) a service for the Merchant Marine and for fishermen, and (6) the "Peace and Progress" station. It is the Moscow output, supplemented by the foreign services from the regional stations, that has the widest coverage and makes use of more than 70 languages, including more for Africa and Asia than any other country and including even, for example, Quechua (for Peru). The "Peace and Progress" output is in German and English for Europe; in Persian, Azerbaijani, and English for Asia; in Arabic and Hebrew for the Middle East; and in Spanish, Portuguese, Creole (for Haiti), and Guaraní (for Paraguay) for Latin America. In 1978 Soviet broadcasting initiated its "World Service in English."

Sweden. Two important developments affecting Sveriges Radio, the Swedish broadcasting corporation, occurred in 1967, embodied in the Broadcasting Law, effective from July 1 of that year. Public authorities and agencies were specifically forbidden to examine programs in advance or to attempt to prevent them from being broadcast; this meant that the government had not even the power of veto. The legal responsibility for any program rests not on the organization, its board of directors, or even the director general but on the program supervisor, and no program may be broadcast against his will. Program content is ultimately controlled by the Radio Council, which supervises both radio and television. The Radio Council may only rule on shows that have already been aired, thus avoiding a role of censorship. The financing is almost entirely from licenses for receiving sets (a small amount of revenue is derived from the international sale and distribution of some radio and television programs), but the proceeds are allocated by Parliament to Sveriges Radio, which produces most of the programs; the Swedish Telecommunications Administration, which transmits them; and the Swedish General Services Administration, which is responsible for the construction of broadcasting facilities. In 1979 Sveriges Radio was broken up into several subsidiary companies: Swedish Radio Company, Swedish Television Company, Swedish Local Radio Company, and Swedish Educational Broadcasting Company. Sveriges Radio became the parent company responsible for long-range planning and development and the distribution of finances to the subsidiaries.

Sweden's television networks

Sweden has three radio and two television networks. A substantial number of stations and transmitters on long waves, medium waves, and shortwaves ensures national coverage of the three radio services, as well as allowing for regional broadcasting. Twenty-four regional stations have a substantial autonomy and their own budgets, but they must negotiate with the heads of the national networks to opt out, with their own regional programs, for up to a total of 25 percent of the network programming. In radio, one network broadcasts spoken-word programs almost exclusively, with some classical music during the day; the second consists of education and light as well as classical music in the evening; and the third, a 24-hour operation, features popular music, news, light entertainment, and regional broadcasts. The two television networks offer a wide variety of features, which include information (17.9 percent), drama and film (13.6 percent), entertainment (13.1 percent), programs for children (10.8 percent), news (9.6 percent), sports (9.5 percent), and education (8.9 percent). Colour television was inaugurated in April 1970.

Sveriges Radio is also responsible for Sweden's external services, the cost of which (as with the cost of educational broadcasts) is separately budgeted and paid for from government funds. The broadcasts are in Swedish and six foreign languages: English, French, German, Spanish, Portuguese, and Russian. They are beamed as appropriate to all parts of the world.

United Kingdom. *British Broadcasting Corporation.* A monopoly until 1954, the BBC operates under a royal charter. It is funded from a fixed-term license fee paid by households with a television set. The BBC has four national radio networks: Radio 1, broadcasting mostly popular music, mainly during the day; Radio 2, primarily transmitting light music, sports, and entertainment; Radio 3, broadcasting mainly classical music and news during the day and cultural programs in the evening; and Radio 4, scheduling spoken word primarily, school programs in the midmorning and early afternoon, and a mixed program in the evenings. The main ingredients of overall output are: 42.9 percent entertainment and music; 21.2 percent classical music; 9.1 percent news and outside broadcasts; 4.8 percent drama; 3.6 percent education; and 2.2 percent features. Some 30 local radio stations have been added to the BBC since 1967. The BBC has two national television services, which together transmit more than 200 hours a week; both have mixed programs that are coordinated to avoid conflicts. The main ingredients are news, documentaries, and information (31 percent); British and foreign films and series (15.5 percent); outside broadcasts, substantially sports, and sports news (14 percent); drama (8 percent); "family" programs and light entertainment (13.5 percent); education (11.1 percent); and religion (2.2 percent).

There is substantial regional activity in both media. Of the six regions in the kingdom that formerly operated with a fair degree of autonomy, only the "national" regions remain for Scotland, Wales, and Northern Ireland. In place of the other three regions, North, Midland, and West and South, there are 20 production centres for both radio and television. Regions broadcast their own programs by opting out of Radio 4 or BBC 1 and using their own section of the corresponding network. Radio Cymru broadcasts in the Welsh language for Wales. There are about 50 local FM (VHF) stations as authorized by the government: these are mostly placed to cover the larger city areas. Many competitive commercial local stations have been set up under the supervision of the Independent Broadcasting Authority.

The BBC is also responsible for the United Kingdom's external services, which are paid for by annual grants-in-aid from the Foreign and Commonwealth Office. Though no longer among the leaders in quantity of output, the

BBC remains among them in terms of penetration. Seventy transmitters, of which 13 are overseas relay stations, provide a shortwave worldwide service and a medium-wave service in many areas, including Europe (from Berlin), Asia, East Africa, and Latin America. Of the weekly output of about 740 hours, roughly one-third in the World Service is in English, and the remainder is in nearly 40 foreign languages, of which the main ones are Arabic (119 hours), French for Europe and Africa (more than 45), Russian and German (55 and 30), Spanish for Spanish-speaking America (28), Polish and Czech (24 and 15), Hungarian (16), and Portuguese for Brazil and Portugal (23).

Independent Broadcasting Authority. The Independent Broadcasting Authority (originally called Independent Television Authority) was established by act of Parliament in 1954; broadcasting began under its control a year later. Although the authority has substantial independence, it does not produce any programs or advertising; these tasks are performed by commercial program companies. These latter, organized on a regional basis, supply all the material broadcast except for news, for which a separate group, Independent Television News, was created, jointly owned and financed by the program companies. TV-am originated in 1983 and operates outside the system with an early morning breakfast-show format. Channel 4 is a wholly owned subsidiary of the International Broadcasting Authority and is funded through compulsory annual subscriptions of Independent Television companies in exchange for advertising rights. It was authorized by the Broadcasting Act of 1980 and began broadcasting in 1982. A separate Welsh fourth channel was authorized at the same time.

Control of the British program companies

The program companies are under a substantial measure of control from the Independent Broadcasting Authority, which is responsible for the appointment of program companies, control of program and advertising output, and its transmission. The authority enforces codes with respect to advertising and violence on the screen. Television companies broadcast throughout the week within their respective areas, except for two that share the London area. The program companies are entirely financed by spot advertising (sponsored programs are forbidden by law) in "natural breaks" in and between programs; they pay a rental to the authority to cover the latter's transmitting and administrative costs and a fiscal levy to the exchequer. The program companies cooperate in a network committee, and a substantial number of the principal programs are broadcast by all companies. The contribution to the network made by each company varies in accordance with its size and resources. The revenue of each company is substantially dependent upon the number of homes with television receivers able to receive the Independent Broadcasting Authority signal in the area it covers, which varies from 32,000 in the Channel Islands to 2,630,000 in the Midlands and 4,300,000 in the London area. So diversified an output makes valueless any percentage analysis of program categories; but the principal types of output, in order of size, are as follows: drama, including telefilm series; light entertainment; children's programs; news, news magazines, features, and documentaries; sports; feature films (British and foreign); education; and religion. The Television Act of 1964 established the present authority situation for 12 years, and its license has since been extended. The authority has also been given the task of supervising local commercial radio (hence its change of name).

The Cable Authority oversees cable television and began operations in 1985.

United States. Broadcasting in the United States is regulated by the Federal Communications Commission, created in 1934, which assigns frequencies and grants licenses. So great is the broadcasting operation in the United States, so many are the stations, both radio and television, and so extensive are the ramifications and links with other industries that it is not possible to produce a summary on the lines of those for countries where broadcasting has been more tightly organized. Some idea of the magnitude of the broadcasting scene is provided by the number of broadcasting stations in operation in the late 1980s, as authorized by the Federal Communications Commission: radio AM, about 5,000; radio FM, about 4,000; educational radio FM, about 1,200; commercial television, about 1,000; educational television, about 300; and television relay stations, more than 4,000. In most categories more stations have been authorized than are operating. The largest proportional increase has been in educational radio FM. Commercial broadcasting on television, as on radio in the past, is dominated by the three great national networks: the American Broadcasting Company, Columbia Broadcasting System, and the National Broadcasting Company. In radio, where the networks are no longer dominant, there is also the Mutual Broadcasting System; the majority of radio stations are as independent of the large networks as of the government, and many of the commercial stations specialize in a single type of output, which may be one or other of various kinds of popular music, classical music, news, or even traffic information. A few are owned by or affiliated with the national networks or with smaller local networks; some even are small local stations offering a basic fare of neighbourhood gossip interspersed with recorded music and spot advertising. After a slump following the major onset of television, radio, even network radio, has again become profitable. In television the three major networks own and operate their own stations in some of the larger cities and substantially control a majority of affiliates.

U.S. noncommercial broadcasting

Noncommercial broadcasting has risen in the United States. The National Association of Educational Broadcasters serves educational stations with transcriptions produced by its members and by other domestic as well as foreign broadcasters. The National Public Radio is also largely educational, supported by donations from foundations and other sources. There are radio stations supported by donations and subscriptions from listeners, in particular the Pacifica group. The Public Broadcasting Service has a loose organization. Its production facilities are not jointly organized, and it makes use of noncommercial stations for its network. Its revenue is uncertain; it received $137,000,000 in 1982 from a congressional appropriation, which must be renewed annually, and the rest from foundations, public contributions, and individual stations.

Another system is community antenna television (CATV), increasingly known as cablevision, or cable TV, originally set up in areas of poor reception or where the choice of receivable television services was poor and cablevision could offer additional choices from elsewhere. By 1964, about 1,000 such systems were in operation. At the time, no one thought of "cablecasting"—*i.e.,* that the cablevision companies should originate their own programs—but in many areas cablecasting has proved a success. Cablevision, transmitted via direct cables connected to each television set, offers viewers a large choice of programs, as well as excellent reception.

Official external services are operated by the Board for International Broadcasting, known as the Voice of America. They are broadcast to all parts of the world and have a number of relay stations overseas. Apart from English, 41 languages are used. There is also the foreign-based Radio Free Europe/Radio Liberty, which broadcasts 1,000 hours a week to eastern Europe and to the Soviet Union. In addition there is the International Broadcast Station, KGEI, offering a shortwave service to Latin America in English, Spanish, and German and in Russian, Byelorussian, Polish, and Ukrainian to Asia; Radio New York World Wide, with a commercial service to Europe and Africa in English and to the Caribbean in English and Spanish; and World International Broadcasters, whose shortwave commercial service with a religious background is broadcast in English to Europe, the Middle East, and North Africa. The United States Armed Forces Radio and Television Service has a network of shortwave stations broadcasting a worldwide service; stations are located in Alaska, Canada, Europe, North Africa, Ethiopia, the Caribbean, East Asia, the Middle East, Antarctica, the North Atlantic, and the Pacific.

Germany's provincial organization

West Germany. Though Germany was one of the first countries to begin radio transmissions (October 1923), the state organization owes nothing to earlier development. It was the occupying powers at the end of World War II that established the present system based on nine state (*Länder*) organizations. All state organizations have a First Radio service on medium wave, supported by FM. Five have a Second, Third, and Fourth Radio service on FM, and the Cologne group has a Fifth. Berlin has an AM channel, as well as FM for each of its three radio services. In many cases, two or more *Länder* organizations cooperate and broadcast simultaneously a single output on one of their FM services. The latter in most cases is broadcast for only three to four hours daily and is substantially, sometimes entirely, devoted to foreign languages for foreign workers in West Germany. The output of the First and Second Radio services is to some extent mixed, but the Second focuses on more serious output.

The First German Television service is nationally coordinated with contributions from each *Land* organization. Each organization broadcasts a substantial amount of regional material for its own audience. The Second Television service is centrally planned and produced, with headquarters in Mainz; schedules are coordinated to give the viewer a maximum choice. Each organization has also a Third Television service for only a few hours a day, often of an educational nature; this service, like the radio Third service, is sometimes produced and simultaneously broadcast by groups of two or three *Länder* organizations. The Federal government has no authority for the control of broadcasting within the West German territory, and legislative and administrative competence for broadcasting rests with the *Länder*. But even the *Land* governments and parliaments are legally barred from intervening beyond a statutory supervision and may not interfere with the basic independence of the broadcasting organizations.

There are two external service organizations. Deutschlandfunk broadcasts to France, the United Kingdom, The Netherlands, and northern and eastern Europe in 12 languages. Deutsche Welle broadcasts in 34 foreign languages to most areas of the world for a total of about 800 hours weekly. (J.A.Ca./Ed.)

Radio systems and equipment

BASIC PHYSICAL PRINCIPLES

Electromagnetic radiation includes light as well as radio waves, and the two have many properties in common. Both are propagated through space in approximately straight lines at a velocity of about 300,000,000 metres (186,000 miles) per second and have amplitudes that vary cyclically with time; that is, they oscillate from zero amplitude to a maximum and back again. The number of times the cycle is repeated in one second is called the frequency (symbolized as f) in cycles per second, and the time taken to complete one cycle is $1/f$ seconds, sometimes called the period. To commemorate the German pioneer Heinrich Hertz, who carried out some of the early radio experiments, the cycle per second is now called a hertz so that a frequency of one cycle per second is written as one hertz (abbreviated Hz). Higher frequencies are abbreviated as shown in Table 3.

Frequency

A radio wave being propagated through space will at any given instant have an amplitude variation along its direction of travel similar to that of its time variation, much like a wave traveling on a body of water. The distance from one wave crest to the next is known as the wavelength.

Wavelength and frequency are related. Dividing the speed of the electromagnetic wave (c) by the wavelength (designated by the Greek letter lambda, λ) gives the frequency: $f = c/\lambda$. Thus a wavelength of 10 metres has a frequency of 300,000,000 divided by 10, or 30,000,000 hertz (30 megahertz). The wavelength of light is much shorter than that of a radio wave. At the centre of the light spectrum the wavelength is about 0.5 micron (0.0000005 metre), or a frequency of 6×10^{14} hertz or 600,000 gigahertz (one gigahertz equals 1,000,000,000 hertz). The maximum frequency in the radio spectrum is usually taken to be about 45 gigahertz, corresponding to a wavelength of about 6.7 millimetres. Radio waves can be generated and used at frequencies lower than 10 kilohertz ($\lambda = 30{,}000$ metres).

Table 3: Frequency Terms and Their Abbreviations

term	cycles per second	abbreviation	equivalent
1 hertz	1	1 Hz	
1 kilohertz	1,000	1 kHz	1,000 Hz
1 megahertz	1,000,000 (10^6)	1 MHz	1,000 kHz
1 gigahertz	1,000,000,000 (10^9)	1 GHz	1,000 MHz

Mechanism of wave propagation. A radio wave is made up of electric and magnetic fields vibrating mutually at right angles to each other in space. When these two fields are operating synchronously in time, they are said to be in time phase; *i.e.*, both reach their maxima and minima together and both go through zero together. As the distance from the source of energy increases, the area over which the electric and magnetic energy is spread is increased, so that the available energy per unit area is decreased. Radio-signal intensity, like light intensity, decreases as the distance from the source increases.

A transmitting antenna is a device that projects the radio-frequency energy generated by a transmitter into space. The antenna can be designed to concentrate the radio energy into a beam like a searchlight and so increase its effectiveness in a given direction (see ELECTRONICS).

Function of transmitting antenna

Frequency bands. The radio-frequency spectrum is divided arbitrarily into a number of bands from very low frequencies to superhigh frequencies (see Table 4). Sections of the spectrum have been allocated to the various users (see Table 5), such as telegraph, telephonic speech, telemetry, and radio and television broadcasting.

Table 4: Frequency Band Designations

frequency designation	frequency range	wavelength range
Very low frequencies (vlf)	3–30 kilohertz	100,000–10,000 m
Low frequencies (lf)	30–300 kilohertz	10,000–1,000 m
Medium frequencies (mf)	300–3,000 kilohertz	1,000–100 m
High frequencies* (hf)	3–30 megahertz	100–10 m
Very high frequencies (vhf)	30–300 megahertz	10–1 m
Ultrahigh frequencies (uhf)	300–3,000 megahertz	1 m–10 cm
Superhigh frequencies (shf)	3–30 gigahertz	10–1 cm

*Also called shortwaves.

The radio-frequency bandwidth is the range of frequencies covered by the modulated radio-frequency signal. The information carried by the signal has a certain bandwidth associated with it, and the carrier must have a channel width at least as great as the information bandwidth. For regular amplitude-modulated (AM) broadcasting the radio-frequency bandwidth must be twice the information-frequency bandwidth. Teleprinter and telex operation requires only a small bandwidth, on the order of 200 hertz, depending on the maximum speed of the pulses forming the information code. Telephonic speech must have high intelligibility, but naturalness (high fidelity) is not of great importance. Tests have shown that the main components of speech lie between about 300 and 3,500 hertz, and telephonic channels carried by radio are therefore normally confined to a bandwidth of about four kilohertz. The smaller the information bandwidth employed the more speech channels can be carried in a given carrier bandwidth and the more economical the system will be.

Young people can hear audio frequencies ranging from about 30 hertz to 18 kilohertz, but as humans grow older, hearing ranges from about 100 hertz to 10 kilohertz. For high-quality (high-fidelity) reproduction of voice or speech, the range should be not less than about 30 hertz (the lowest frequency of a large organ pipe) to 15 kilohertz (piccolo, cymbal, triangle). Acceptable audio quality under certain circumstances may be achieved with a bandwidth as small as five kilohertz, as in AM radio; a much larger bandwidth is needed for transmitting a moving picture because it is necessary to convey the overall average light content of a picture as well as the picture detail. The average light content requires frequencies as low as 20 hertz to be transmitted, and picture detail demands frequencies up to five megahertz for a standard television picture.

Minimum acceptable bandwidths

Modulators and demodulators. A carrier wave is a radio-frequency wave that carries information. The information is attached to the carrier wave by means

Table 5: Utilization of Radio-Frequency Spectrum

very low frequencies (vlf)		time signals, standard frequencies
	30 kHz	
low frequencies (lf)		fixed, maritime mobile, navigational, radio broadcasting
	300 kHz	
medium frequencies (mf)		land, maritime mobile, radio broadcasting
	3 MHz	
high frequencies (hf)		fixed, mobile, maritime and aeronautical mobile, radio broadcasting, amateur
	30 MHz	
very high frequencies (vhf)		fixed, mobile, maritime and aeronautical mobile, amateur, radio and television broadcasting, radio navigation
	300 MHz	
ultrahigh frequencies (uhf)		fixed, mobile, maritime and aeronautical mobile, amateur, television broadcasting, radio location and navigation, meterological, space communication
	3 GHz	
superhigh frequencies (shf)		fixed, mobile, radio location and navigation, space and satellite communication
	30 GHz	

of a modulation process that involves the variation of one of the carrier-frequency characteristics, such as its amplitude, its frequency, or its duration. (All of these processes are discussed in greater detail in the article TELECOMMUNICATIONS SYSTEMS.)

In amplitude modulation the information signal varies the amplitude of the carrier wave, a process that produces a band of frequencies known as sidebands on each side of the carrier frequency. These sidebands (a pair to each modulation frequency) cover a range of frequencies equal to the sum and difference between the carrier frequency and the information signal.

Frequency modulation involves varying the frequency (the number of times the wave passes through a complete cycle in a given period of time, measured as cycles per second) of the carrier in accordance with the amplitude of the information signal. The amplitude of the carrier wave is unaffected by the variation; only its frequency changes. Frequency modulation produces more (often many more) than one pair of side frequencies for each modulation frequency.

The variation of carrier frequency is known as the frequency deviation, and for very-high-frequency broadcasting it can reach ± 75 kilohertz. The greater the frequency deviation the greater is the effective modulation. Though theoretically its maximum value need not be limited to 75 kilohertz, any increase beyond this value requires a wider channel, which adds to the cost of reception and reduces the number of transmitters that can be accommodated in the band. The total channel width is approximately twice the sum of the maximum deviation frequency and modulating frequency. If channel width is restricted in either transmitter or receiver circuits, distortion of the information signal occurs.

A radio broadcast normally consists of only one information signal. The listener hears what he would hear at the microphone position if only one of his ears was functioning; *i.e.*, it is a monaural system. In such a system it is not possible to gain any impression of the position of the instrument groupings in an orchestra, nor can lateral movement be indicated, though movement toward or away from the microphone is conveyed by a change in sound volume.

Stereophonic broadcasting requires two microphones, one to collect sounds from the left and one from the right; the two sets of information must be separable in the receiver pnd be fed to loudspeakers on the left and on the right at the listening position. For high-fidelity reproduction the full audio range up to 15 kilohertz is transmitted; this can only be achieved satisfactorily at very high frequencies with frequency modulation. The broadcast signal is received on monaural receivers by making one set of information the sum of left and right signals (L + R). The other set of information is the difference of left and right (L − R). Summation of the two sets of information at the receiver output recovers the left (L) signal and subtraction recovers the right (R) signal.

Pulse code modulation

Another system of modulation switches the carrier on and off in pulses, the duration or position of the pulse being determined by the information signal. This system of pulse-coded modulation can provide better protection from noise, and a number of separate speech channels can be combined by allocating specified groups of pulses for each information channel and then interleaving these pulses in a process called time division multiplex. To accomplish this, a comparatively wide transmission channel is needed, and the carrier must be an ultrahigh or superhigh frequency.

The ionosphere. An English mathematician, Oliver Heaviside, and a U.S. electrical engineer, Arthur Edwin Kennelly, almost simultaneously predicted in 1902 that radio waves, which normally travel in straight lines, are returned to Earth when projected skyward because electrified (ionized) layers of air above the Earth (the ionosphere) reflect or refract (bend) them back to Earth, thus extending the range of a transmitter far beyound line of sight. In 1923 the suggestion was proved to be accurate when pulses of radio energy were transmitted vertically upward and returning pulses were received back from the reflecting layer. By measuring the time between the outgoing and returning pulses, it was possible to estimate the height and number of layers. Three layers can normally be distinguished at distances from 50 to about 400 kilometres (30 to 250 miles) above the Earth's surface. The layers result from a breakdown of gas atoms into positively charged ions and free electrons caused by energy radiated from the Sun. The electrons maintain a separate existence in the lower layers for as long as the Sun's energy is being received, and in the upper layers some can remain free throughout the hours of darkness.

The three layers are designated D, E, and F. The D layer is approximately 80 kilometres (50 miles) high and exists only during daylight hours. Because it absorbs medium frequencies and the lower frequencies of the shortwave bands, it limits the range of such stations during daylight. The E layer, about 110 kilometres (68 miles) high, maintains its reflectivity for four or five hours after the Sun sets and so extends the range of such stations to as much as 1,000 kilometres (620 miles). This layer also serves as a good reflector of shortwaves during the day and into the night, until its reflectivity drops.

Most important of the three layers is the F layer, which has considerable power to reflect the higher frequencies. During the day it often splits into two layers (F_1 and F_2) at about 200 and 400 kilometres (125 and 250 miles), but at night only one layer is generally present at a height of about 300 kilometres (190 miles).

Radio noise, fading, and interference. Any sudden discharge of electrical energy, like that of lightning, produces transient (short-duration) radio-frequency waves, which are picked up by antennas. These packets of radio-frequency energy produce the crackle heard on an amplitude-modulated radio receiver when an electrical storm is nearby and may be classed as natural noise.

Switching of high-voltage power lines can produce similar effects; the lines help to carry the noise-producing signals over long distances. Local switching of lights and electrical machinery can also produce the familiar crackle when the receiver is close to the noise-producing source. These sources are classed as man-made noise.

Filters in noise production

Generally noise of both types decreases as the frequency is increased. An exception is automobile ignition noise, which produces maximum effect in the very-high-frequency range, causing a sound in nearby loudspeakers every time a spark plug fires. Many countries have legislation requiring the suppression of man-made noise by means of filters that reduce the amount of radio-frequency energy released at the source. Metallic shielding of leads to and from the noise source curtails the radiated interference. It is also possible to install various noise-reducing devices at the input to radio receivers.

Noise is also caused by irregularities in the flow of electrons in metals, transistors, and electron tubes. This source of noise ultimately limits the maximum useful signal amplification that can be provided by a receiver. Noise due to the random movement of electrons causes a hiss in the loudspeaker. Radio noise can also be picked up from outer space as a hiss similar to random electron noise.

Fading of a signal, on the other hand, is due to variation in the propagation characteristics of the signal path or paths. This is particularly true when propagation depends on reflection from the ionosphere as it does for shortwaves. Propagation of waves in the very-high-frequency range and above, which penetrate the ionosphere, can be affected by temperature changes in the stratosphere, that part of the atmosphere up to about 15 kilometres (nine miles) from the Earth's surface. The fading effect can be greatly reduced at the receiver loudspeaker by various electronic controls, such as automatic gain control.

The phenomenon of interference occurs when an undesired signal overlaps the channel reserved for the desired signal. By interaction with the desired carrier, the undesired information may cause speech to become unintelligible. Countermeasures include narrowing the desired channel, thus losing some information but preventing overlap, and using a directional antenna to discriminate against the undesired transmission.

DEVELOPMENT OF RADIO TECHNOLOGY

Maxwell's prediction. Early in the 19th century, Michael Faraday, an English physicist, demonstrated that an electric current can produce a local magnetic field and that the energy in this field will return to the circuit when the current is stopped or changed. James Clerk Maxwell, professor of experimental physics at Cambridge, in 1864 proved mathematically that any electrical disturbance could produce an effect at a considerable distance from the point at which it occurred and predicted that electromagnetic energy could travel outward from a source as waves moving at the speed of light.

Maxwell's prediction verified

Hertz: radio-wave experiments. At the time of Maxwell's prediction there were no known means of propagating or detecting the presence of electromagnetic waves in space. It was not until about 1888 that Maxwell's theory was tested by Heinrich Hertz, who demonstrated that Maxwell's predictions were true at least over short distances by installing a spark gap (two conductors separated by a short gap) at the centre of a parabolic metal mirror. A wire ring connected to another spark gap was placed about five feet (1.5 metres) away at the focus of another parabolic collector in line with the first. A spark jumping across the first gap caused a smaller spark to jump across the gap in the ring five feet away. Hertz showed that the waves travelled in straight lines and that they could be reflected by a metal sheet just as light waves are reflected by a mirror.

Marconi's development of wireless telegraphy. The Italian physicist Guglielmo Marconi, whose main genius was in his perseverance and refusal to accept expert opinion, repeated Hertz's experiments and eventually succeeded in getting secondary sparks over a distance of 30 feet (nine metres). In his experiment he attached one side of the primary spark gap to an elevated wire (in effect, an antenna) and the other to Earth, with a similar arrangement for the secondary gap at the receiving point. The distance between transmitter and receiver was gradually increased first to 300 yards (275 metres), then to two miles (three kilometres), then across the English Channel. Finally, in 1901, Marconi bridged the Atlantic when the letter *s* in Morse code travelled from Poldhu, Cornwall, to St. John's, Newfoundland, a distance of nearly 2,000 miles (3,200 kilometres). For this distance, Marconi replaced the secondary-spark detector with a device known as a coherer, which had been invented by a French electrical engineer, Edouard Branly, in 1890. Branly's detector consisted of a tube filled with iron filings that coalesced, or "cohered," when a radio-frequency voltage was applied to the ends of the tube. The cohesion of the iron filings allowed the passage of current from an auxiliary power supply to operate a relay that reproduced the Morse signals. The coherer had to be regularly tapped to separate the filings and prepare them to react to the next radio-frequency signal.

The Fleming diode and De Forest Audion. The next major event was the discovery that an electrode operating at a positive voltage inside the evacuated envelope of a heated filament lamp would carry a current. The American inventor Thomas A. Edison had noted that the bulb of such a lamp blackened near the positive electrode, but it was Sir John Ambrose Fleming, professor of electrical engineering at Imperial College, London, who explored the phenomenon and in 1904 discovered the one-directional current effect between a positively biased electrode, which he called the anode, and the heated filament; the electrons flowed from filament to anode only. Fleming called the device a diode because it contained two electrodes, the anode and the heated filament; he noted that when an alternating current was applied to the diode, only the positive halves of the waves were passed—that is, the wave was rectified (changed from alternating to direct current). The diode could also be used to detect radio-frequency signals since it suppressed half the radio-frequency wave and produced a pulsed direct current corresponding to the on and off of the Morse code transmitted signals. Fleming's discovery was the first step to the amplifier tube that in the early part of the 20th century revolutionized radio communication.

Fleming failed to appreciate the possibilities he had opened up and it was the American inventor Lee De Forest who in 1906 conceived the idea of interposing

Invention of the Audion

an open-meshed grid between the heated filament and positively biased anode, or plate, to control the flow of electrons. De Forest called his invention an Audion. With it he could obtain a large voltage change at the plate for a small voltage change on the grid electrode. This was a discovery of major importance because it made it possible to amplify the radio-frequency signal picked up by the antenna before application to the receiver detector; thus, much weaker signals could be utilized than had previously been possible.

Early research by commercial companies. The first commercial company to be incorporated for the manufacture of radio apparatus was the Wireless Telegraph and Signal Company, Ltd. (England) in July 1897 (later changed to Marconi's Wireless Telegraph Company, Ltd.); other countries soon showed an interest in the commercial exploitation of radio.

Among the major developments of the first two decades of the 20th century was De Forest's discovery in 1912 of the oscillating properties of his Audion tube, a discovery that led to the replacement of the spark transmitter by an electronic tube oscillator that could generate much purer radio waves of relatively stable frequency. By 1910, radio messages between land stations and ships had become commonplace, and in that year the first air-to-ground radio contact was established from an aircraft. A landmark transmission came in 1918, when a radiotelegraph message from the Marconi long-wave station at Caernarvon, in Wales, was received in Australia, over a distance of 11,000 miles (17,700 kilometres).

Though early experiments had shown that speech could be transmitted by radio, the first significant demonstration was not made until 1915 when the American Telephone & Telegraph Company successfully transmitted speech signals from west to east across the Atlantic between Arlington, Virginia, and Paris. A year later, a radiotelephone message was conveyed to an aircraft flying near Brooklands (England) airfield. In 1919 a Marconi engineer spoke across the Atlantic in the reverse direction from Ballybunion, Ireland, to the United States.

From 1920 onward radio made phenomenal progress through research activities in Europe, North America, and Asia. The invention of the electron tube and later the transistor (1948) made possible remarkable developments.

Hand wiring. In the early days of radio and up to the end of World War II, radio receivers consisted of resistors, capacitors, inductors (coils), and electronic tubes joined together by wires with coloured insulation. A colour code, whereby a particular colour was assigned to a particular circuit connection, such as black leads for filaments, green for grid, was adopted throughout the world to facilitate manufacture and the tracing of faults. Later, wires cut to the right length were laced together into a harness to speed assembly. Plugs and sockets were employed for connecting one major part with another. Printed circuit wiring, developed during the 1940s, eliminated much of the hand work and produced important manufacturing economies.

Printed circuits. With printed wiring, the layout of the circuit is planned with component size and position in mind, and connections are made by suitably shaped copper strip or foil bonded to an insulating board or substrate. An extension of this technique was the printed component; resistors, capacitors, and low value inductors became a part of the printing process.

The development of the transistor simplified the exploitation of printed circuitry by eliminating one of the bulkiest components, the vacuum tube. Further development led to the manufacture of the integrated circuit in the 1960s. Compact circuits of this type can perform a multiplicity of tasks such as amplification and switching. They are widely used in computers where space is at a premium. Integrated-circuit amplifiers are likely to become more important because of their ability to amplify very high frequencies.

The size of a portable receiver constructed from microminiature circuits is now dictated almost entirely by the loudspeaker and the quality of reproduction required. The smaller the loudspeaker the lower the power it can accept and the less the output of low audio frequencies.

RADIO CIRCUITRY

Components. The basic operating principles of the major circuitry and active and passive components used in radio are described in the article ELECTRONICS. In this section, only enough description is included to permit the reader to understand the applications to radio circuitry.

Primary function of the electron tube

Active devices: vacuum tubes and transistors. An electron tube or transistor, designated an active element, functions basically as an amplifier, and its output is essentially an amplified copy of the original input signal. The simplest amplifying electron tube is the triode, consisting of a cathode coated with material that provides a copious supply of electrons when heated, an open-mesh grid allowing electrons to pass through but controlling their flow, and a plate (anode) to collect the electrons. The plate is maintained at a positive voltage with respect to the cathode in order to attract the electrons; the grid usually has a small negative voltage so that it does not collect electrons but does control their flow to the plate. The output voltage is usually many times greater than the input voltage to the grid. The tube must be pumped to a high degree of vacuum, or the plate current flow is erratic.

Other electrodes, also in the form of open-mesh grids, may be included in the tube to perform various special functions. An example is the four-electrode tube known as the tetrode, in which an open-mesh grid (screen grid) maintained at a positive voltage is placed between plate and control grid. This reduces the effect of plate voltage on electron flow and increases the amplifying property of the tube. Introduction of a third grid, known as a suppressor grid, produces the pentode (five-electrode tube), which can provide even greater amplification.

The transistor, which has largely replaced the electron tube as the active element in low-voltage electronic circuits, is made from semiconductor materials—that is, substances that are neither good conductors nor good insulators. Two common semiconductor materials are germanium and silicon, to which small amounts of impurities such as indium, gallium, arsenic, or phosphorus are added to impart electrical charges to them. Arsenic and phosphorus, for example, provide extra negative charges, giving *n*-type (signifying excess negative charges) material; indium or gallium yield a shortage of electrons or an excess of positive charges or holes, giving *p*-type (signifying excess positive charges) material.

A transistor is a sandwich of semiconductor materials with the same impurity in the two outer layers and a different impurity in the centre layer providing current carriers of opposite charge to those produced by the outer layers.

If the outer layers are reservoirs of positively charged current carriers (*p* type) and the centre layer provides an excess of electrons (*n* type), the transistor is known as a *p–n–p* (positive–negative–positive carriers) type. If the *p* and *n* layers are reversed, the transistor is an *n–p–n* type. The two outer layers are termed the emitter and collector, and the centre layer is known as the base.

A transistor is an amplifier of current; the vacuum tube, in contrast, is an amplifier of voltage. The transistor produces an adequate supply of current carriers (electrons and holes) at room temperature and does not require a heated cathode as does the vacuum tube. Thus the power required from the power supply is much reduced, less heat is produced, and the transistors and their circuitry can be packed into a smaller container and takes up less space. Transistors are also physically much smaller than comparable electron tubes. Thus the transistorized portable radio can fit in a pocket—a distinct advantage over the cumbersome tube radio it has replaced.

In its early form the transistor was capable of amplifying only comparatively low frequencies because the exchange of electrons and positive charges across the sandwich was slow. Modern techniques however, have overcome this difficulty so that amplification up to frequencies over 1,000 megahertz is commonplace.

Tuned circuits and the superheterodyne principle. For information (voice, music, television) to be transmitted, it must be attached to a radio-frequency carrier wave, which is then transmitted in a given frequency channel. The carrier wave and information can be picked up by a receiver tuned to this channel. The process by which the information is attached to the carrier wave is modulation. Modulated carriers are isolated in their separate slots or channels; if transmitters are geographically close to each other, they must not use the same channel or overlap each other's channels. If such overlap occurs, serious interference results—two radio programs may be heard simultaneously or one may form a distorted background to the other.

Modulation

In most modern radio receivers, reception is based on the superheterodyne principle. The incoming radio frequency is mixed (heterodyned) with the output of an oscillator the frequency of which is adjusted so that the difference between it and the incoming signal is constant; the result is the intermediate frequency. Amplification is thereafter carried out at this intermediate frequency. Both preliminary selection of the incoming frequency and adjustment of the local oscillator frequency are accomplished by variable tuned circuits consisting of inductance and capacitance. Tuning may be accomplished by varying the capacitance, which consists of interleaved metal plates separated by air spaces with one set of plates movable. Another method of tuning involves varying the inductance by insertion or withdrawal of an iron dust or ferrite core in a cylindrical coil of copper wire. To simplify the tuning procedure, the variable elements of all stages requiring tuning are ganged together and coupled to a tuning knob. The intermediate-frequency amplifier stages always operate at the same frequency and so require tuning only when the receiver is manufactured or serviced. The intermediate frequency is usually around 455 kilohertz for AM receivers, 10.7 megahertz for FM receivers, and 38 megahertz for television receivers. Most of the amplification in a radio receiver is carried out in the intermediate-frequency stages, and most of the selectivity (ability to separate adjacent stations) is obtained in these stages.

Oscillators. A self-oscillating circuit consists of a vacuum tube or transistor, a tuned circuit, and some form of positive feedback (energy fed from the output back to the input in such a way as to increase the input). Since both tubes and transistors can function as amplifiers, they can also function as oscillators. For receiver circuits, adequate oscillator stability can be obtained with conventional tuned circuits, but the transmitter oscillator must be highly stable, and a circuit made up of inductance and capacitance, tuned to the desired frequency, is not sufficiently stable. A piezoelectric crystal oscillator (a device that vibrates or oscillates at a given frequency emitting radio waves when voltage is applied to it) or its equivalent is ordinarily used.

Amplifiers. Amplifiers may be classified in a number of different ways: according to bandwidth (narrow or wide); frequency range (audio, intermediate, or radio frequency); or output parameter requirement (voltage or power).

Wide-band radio-frequency amplifiers are not needed for audio signals unless a frequency-modulated system is used. Amplitude-modulated signals for sound broadcasting should have a radio-frequency bandwidth of ±10 kilohertz though on medium waves it is often limited to about ±5 kilohertz (total bandwidth of 10 kilohertz). High-quality frequency-modulated audio needs a bandwidth of about ±100 kilohertz.

Audio-frequency amplifiers present few design problems, and negative feedback of the output into the input can overcome distortion problems. Radio-frequency amplifiers, which can be tuned, suffer from variation of selectivity (ability to separate adjacent stations) and gain (amplification) over the tuning range. Selectivity tends to broaden and gain to increase as capacitance is decreased, and instability can be troublesome at the highest tuning frequency. Intermediate-frequency amplifiers do not suffer from these defects since the tuning frequency is fixed.

Use of negative feedback to minimize distortion

The main problem with radio-frequency amplifiers in receivers is the possibility of cross modulation—that is, the mixing of two information channels, which can occur if an undesired modulated signal enters the radio-frequency input together with the desired signal.

Antennas. The antenna is an essential part of a radio transmission and reception system. Its purpose at the transmitter is to project electromagnetic energy into space and at the receiver to extract energy from the travelling electromagnetic wave produced from the transmitter antenna.

The size of the antenna relative to the wavelength of the electromagnetic radiation is important. The wavelength of medium waves is about 300 metres (1,000 feet), and a vertical transmitting mast or self-supporting tower 150 to 210 metres (490 to 690 feet) high may be used with a high-power transmitter (200 kilowatts or more). An equally tall receiving antenna would be ideal but impractical. A vertical rod or suspended wire about six metres (20 feet) long is a workable solution. If the transmitting antenna is vertical, the receiving antenna must also be vertical; if the former is horizontal the receiving antenna must be horizontal. This rule applies at all radio frequencies except shortwaves because the plane of their electromagnetic field can be twisted in its passage through the ionosphere, and a vertical shortwave antenna may pick up a good signal from a horizontal transmitting antenna. The antenna system becomes progressively smaller as the transmitting frequency is increased, and at ultrahigh frequencies (300 megahertz or more) the individual antenna may be only about 50 centimetres (20 inches) long.

For normal amplitude-modulated broadcasting, the receiver antenna may be composed of a fairly short coil of wire wound on a powdered iron or ferrite core. This type of design permits adequate signal pickup with a very small antenna which may be located in a small space, a necessity for small, battery-operated portable receivers.

Antennas may have their directional characteristics modified by employing multiple elements. Thus an antenna may be omnidirectional (transmitting in all directions) horizontally but narrowly beamed vertically, or it may be bidirectional (transmitting in two directions) in a figure eight pattern with two main directions of energy projection at 180°. It may be unidirectional, having energy projected to one side, or the energy may be concentrated in a relatively narrow beam both horizontally and vertically.

Directional characteristics

In point-to-point communication, as from one network centre to another, highly directional antennas are used. Local broadcasting uses an omnidirectional antenna, radiating equally in all directions, except in such special cases as a coastal location or proximity to a neighbouring transmitter.

Broadcasting planned to serve distant areas, employing shortwaves and depending on reflection from the ionosphere, normally uses a relatively narrow beam of energy projected skyward at an angle from 5° to 10° to the horizontal. A reflecting curtain is placed behind the antenna to prevent loss of energy in the reverse direction. The beam is divergent (spreads out) so that after two or three reflections between ionosphere and Earth it covers a relatively large area.

Transmission lines. The lines that carry radio waves from the radio transmitter to the antenna are known as transmission lines; their purpose is to convey radio-frequency energy with minimum heating and radiation loss. Heating losses are reduced by conductors of adequate size. Only the outer layers of the conductor carry radio-frequency current.

Concepts of selectivity and sensitivity. Radio-frequency communication requires the receiver to reject all but the desired signal. Were the number of frequency channels equal to the demand, each channel could be given its correct width in the tuning stages of a receiver. Thus, for audio broadcasting each carrier channel should be 20 kilohertz wide to accommodate both side bands, and each transmission carrier should be 20 kilohertz, separated from those on either side. In much of the world, the medium-wave and shortwave bands are in such demand that transmitters must share the same channel and channels thus must overlap. Though efforts have been made to arrange sharing by geographically separated transmitters, the congestion has forced receiver manufacturers to reduce the receiver bandwidth to about eight kilohertz (±four kilohertz).

Very-high-frequency transmissions can generally be received at full bandwidth because their signals are confined to line of sight and are, in effect, local-station signals to the receiver. Frequency-modulated transmissions must be received on full bandwidth (about 200 kilohertz) if serious distortion is to be avoided on loud programs.

Effect of noise on sensitivity

Receiver sensitivity is the ability of a receiver to pick up weak signals. Though a communication receiver should always have a high sensitivity, there is a maximum determined by the noise generated inside the receiver itself. Little value is gained by increasing sensitivity if noise at the receiver output is already considerable and comparable with desired signal output. Normally, radio broadcasting systems operate with the signal voltage at least 10 to 50 times greater than the noise. To take full advantage of high sensitivity, receiving antennas for communications links are usually located in an area where there is little man-made noise. A receiver intended only for local-station reception can have a much lower sensitivity than a shortwave receiver intended for picking up signals from the other side of the world. (K.R.S./Ed.)

Television systems and equipment

The purpose of a television system is to extend the senses of vision and hearing beyond their natural limits. Television systems are designed, therefore, to embrace the essential capabilities of these senses, with appropriate compromises between the quality of the reproduction and the costs involved. The aspects of natural vision that must be considered in a television system include the ability of the human eye to distinguish the brightness, colours, details, sizes, shapes, and positions of the objects in the scene before it. The aspects of hearing include the ability of the ear to distinguish the pitch, loudness, and distribution of sounds. The television system must also be designed to override, within reasonable limits, the effects of interference and to minimize visual and aural distortions in the transmission and reproduction processes. The particular compromises adopted for public television service are embodied in the television standards adopted and enforced in each country by the government agency responsible for broadcasting.

How television differs from the human eye

Television technology deals with the fact that human vision employs many hundreds of thousands of separate electrical circuits, in the optic nerve from the retina to the brain, to convey simultaneously in two dimensions the whole content of the scene on which the eye is focussed, whereas in electrical communications it is feasible to employ only one such circuit (*i.e.,* the broadcast channel) to connect the transmitter and the receiver. This fundamental disparity is overcome in television by a process of image analysis and synthesis, whereby the scene to be televised is first translated into an electrical image, and the latter is then broken up into an orderly sequence of electrical impulses that are sent over the channel one after the other. At the receiver the impulses are translated back into a corresponding sequence of lights and shadows, and these are reassembled in their correct positions on the viewing screen.

This sequential reproduction of visual images is feasible only because the visual sense displays persistence; that is, the brain retains the impression of illumination for about 0.1 second after the source of light is removed from the eye. If, therefore, the process of image synthesis occurs within less than 0.1 second, the eye is unaware that the picture is being reassembled piecemeal, and it appears as if the whole surface of the viewing screen were continuously illuminated. By the same token, it is then possible to re-create more than ten complete pictures per second and to simulate thereby the motion of the scene so that it appears to be continuous.

In practice, to depict rapid motion smoothly, it is customary to transmit from 25 to 30 complete pictures per second. To provide detail sufficient to accommodate a wide range of subject matter, each picture is analyzed into 300,000 or more elementary details. This analysis implies that the rate at which these details are transmitted over the television system exceeds 4,000,000 per second. To provide a system suitable for public use and also capable of such speed has required the full resources of modern electronic technology.

DEVELOPMENT OF TELEVISION SYSTEMS

Mechanical systems. Early ideas for the realization of television assumed the transmission of every picture element simultaneously, each over a separate circuit (as, for example, a system suggested by George Carey of Boston in 1875); but in about 1880 the important principle—subsequently adopted in all forms of television—of rapidly scanning each element in the picture in succession, line by line and frame by frame, with reliance on persistence of human vision, was proposed, notably by W.E. Sawyer in the United States and Maurice Leblanc in France. This established the possibility of using only a single wire or channel for transmission.

Nipkow system

In 1873 the photoconductive properties of the element selenium were discovered; that is, the fact that its electrical conduction varied with the amount of illumination. This appeared to provide an important clue to the secret of practical television and led in 1884 to a patent by Paul Nipkow in Germany of a complete television system. The distinctive feature of Nipkow's system was the spirally apertured rotating disk that provided, at both sending and receiving ends, a simple and effective method of image scanning. Until the advent of electronic scanning, all workable television systems depended on some form or variation (*e.g.,* mirror drums, lensed disks, etc.) of the mechanical sequential scanning method exemplified by the Nipkow disk.

As illustrated in Figure 1, the image to be televised is focussed on a rotating disk having square apertures arranged in a spiral. As the disk rotates, the outermost aperture traces out a line across the top of the image, and the light passing through the aperture varies in direct proportion to the light and shade (*i.e.,* brightness values) of that line of the image as it is traversed by the aperture. When the outermost aperture has passed over the image, the next inner aperture traces out another line, parallel to and immediately below the line just traced. The changes in the light passing through the second aperture represent, in sequence, the brightness values present in the image along this second line. As the disk continues to rotate, successive lines are traced out, one beneath the other, until the whole area of the image has been explored, one line at a time. The process is repeated with each rotation of the disk; the more apertures, and hence lines, the greater the detail that can be analyzed.

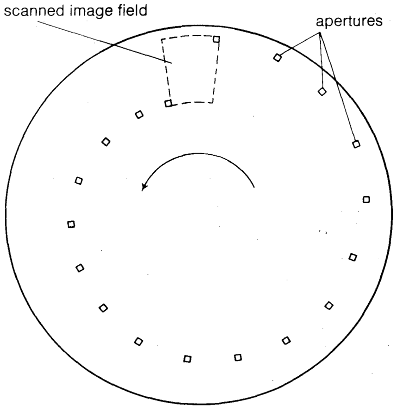

Figure 1: Nipkow disk for an 18-line picture.

In this way the detail of the whole image is sequentially explored in an orderly manner. The light passing through the apertures enters a photoelectric cell that translates the sequence of light values into a corresponding sequence of electric values. These are transmitted over a single circuit to the receiver, where the electrical impulses cause light to be produced by a lamp (such as a gas-discharge lamp) capable of reproducing the sequence of light values. The light from the lamp is projected onto the surface of a disk similar to that at the transmitter. This disk must rotate in precise synchronism, and, by a scanning process the reverse of that already described, the brightness values are reassembled in their proper positions and the original image is reproduced. Provided the rotation is at sufficient speed, persistence of vision enables the eye to see the image as a whole rather than as a series of moving points. The need for exact synchronism between camera and receiver scanning speeds is fundamental not only to the mechanical system but to every television system.

Selenium, when used as the photosensitive material for the photoelectric cell, had a serious handicap: its response to changes of light was very slow. Researches in Germany resulted (1913) in a potassium hydride-coated cell with both an improved sensitivity and the ability to follow rapid changes of light. This made possible for the first time a practical working system.

In 1897 K.F. Braun of Germany had introduced a cathode-ray tube with a fluorescent screen; that is, a screen that produced visible light when struck by a beam of electrons. The Russian scientist Boris Rosing in 1907 suggested its use in the receiver of a television system that, at the camera end, made use of mirror-drum scanning. Rosing succeeded in transmitting and reproducing some crude geometrical patterns and so ranks as an important pioneer.

In 1904 the English physicist J.A. Fleming invented the two-electrode valve; the American inventor Lee De Forest added the grid in 1906 and made amplification possible, another essential step toward practical television.

Swinton's proposals

In 1908 came a most remarkable contribution by the Scottish electrical engineer A.A. Campbell Swinton, who outlined a method that, in all essentials, is the basis of modern television. Lack of amplifiers and other difficulties confined this to what Swinton called "an idea only," but he clarified and elaborated it in an address to the Röntgen Society of London in 1911. He proposed, in essence, the use of cathode-ray tubes, magnetically deflected, at both camera and receiver. In the former was a mosaic screen of photoelectric elements; the image of the scene to be transmitted was focussed onto this screen, the back of which was then discharged by a cathode-ray beam tracing out a line-by-line scanning sequence. Swinton's brilliant ideas were too advanced for early application, and it was left to others to put them into practice many years later.

Meanwhile, experimenters in Europe and the United States were trying to make a less ambitious beginning. The neon gas-discharge lamp, produced by D.M. Moore of the United States in 1917, made it possible to vary the light intensity at the receiver by varying the electrical input to the neon lamp, thus in effect producing modulated light. It was adopted by J.L. Baird in England and C.F. Jenkins in the United States, both of whom in 1923 began experimenting with mechanical methods using the Nipkow principle. In 1926 Baird gave the first demonstration of true television by electrically transmitting moving pictures in halftones. These pictures were formed of only 30 lines, repeating approximately 10 times per second. The results, though inevitably crude—flickering badly and with a dim receiver screen a few inches high—nevertheless became the start of television as a practical technology and did much to stimulate further research, while also forming the basis of some experimental broadcasting in England between 1929 and 1935.

Electronic systems. Mechanical systems lacked sensitivity, as became progressively manifest with attempts to increase the number of lines and thus the degree of definition of the pictures. Swinton and others had pointed out that television pictures, for good quality and definition on a screen of reasonable size, would need to be analyzed into at least 100,000 and preferably 200,000 elements. Since the number of elements is approximately equal to the square of the number of lines, it can be seen that any system using 30 or even 100 lines would be inadequate—300 being more nearly the minimum. Although mechanical systems were with difficulty made to operate on 200 and more lines, thought increasingly turned toward the greater potential of electronic methods. A most important landmark was V.K. Zworykin's patent, first filed in 1923, for the iconoscope camera tube. Later, he constructed such a tube; and by 1932 the Radio Corporation of America (RCA), with an improved cathode-ray tube for the receiver, demonstrated all-electronic television (initially on 120 lines), so proving the soundness of Swinton's theoretical ideas. The compactness and convenience of the electronic camera were remarkable, and its sensitivity, greatly aided by the unique "storage" feature of the iconoscope, was comparable with motion-picture cameras of the time.

Continuing work on electronic systems was greatly stimulated. In the United States the development was mainly carried out in the RCA laboratories; very soon the number of scanning lines was increased to 343, and other improvements followed rapidly. German investigators also were active, especially in the development of high-vacuum cathode-ray tubes. By 1935 a regular broadcasting service had begun in Germany, though with medium definition only—180 lines. In The Netherlands, too, the Philips Laboratories took up television research.

In Great Britain, Electric and Musical Industries (EMI) set up in 1931 a television research group under Isaac Shoenberg. The team produced a complete and practical system, including all of the complex electronics surrounding the camera and receiver tubes, as well as the intervening control and amplifying circuits. Shoenberg saw the need to establish a system that would endure for many years, since any subsequent changes in basic standards—particularly the number of scanning lines and their repetition rate—could give rise to severe technical or economic problems. He therefore proposed the use of 405 lines with 50 frames per second, and interlaced scanning to give 25 pictures per second without flicker. The government authorized the British Broadcasting Corporation to adopt these standards, as well as the complete EMI system. Initially, and for a short time only, the system was under comparison with alternate broadcasts from a 240-line, 25-picture system developed by the Baird company. The latter employed mechanical scanning methods in the camera and suffered from lack of sensitivity as well as other limitations.

United States and European standards

By the mid-1930s electronic television was fast advancing in all its aspects. Important questions were the settling of basic standards (number of lines and frames per second) before the introduction of public broadcasting services in the United States and elsewhere, though these questions were not everywhere fully resolved until about 1951. The United States soon adopted a picture repetition rate of 30 per second, while in Europe it became 25. These two standards have been perpetuated, and all the countries of the world use one or the other, though technical advances have now obviated the original need for disparity. The arguments in relation to the number of lines were based on the need for an effective compromise between, on the one hand, adequate picture definition, and, on the other, a frequency bandwidth that could be technically and economically acceptable. World standardization has never been achieved, though for new television services all countries are adopting one of two main standards, namely, 525 lines per picture at 30 pictures per second—the United States standard—and 625 lines at 25 pictures per second, usually known as the European standard. Complications arise when programs are transmitted between countries using different standards.

Technical advances have been continuous, particularly the great improvements in camera tubes (*e.g.,* the image orthicon and the Vidicon, the latter of which was at last able effectively to exploit the photoconductivity principle) that were made from 1945 onward. By the early 1950s technology had progressed so far, and television had become so widely established, that the time was ripe to tackle in earnest the problem of creating television images in natural colours.

The development of colour television. Colour television was by no means a new idea, since its attraction and possibility early engaged the imaginations of inventors. A German patent in 1904 contained the earliest proposal, while in 1925 Zworykin filed a patent disclosure for an all-electronic colour television system. The first practical demonstration of television in colour was given by Baird in 1928; he used mechanical scanning with a Nipkow disk having three spirals of 30 apertures, one spiral for each primary colour in sequence. The light source at the receiver was composed of two gas-discharge tubes, one of mercury vapour and helium for the green and blue colours, and a neon tube for red. In 1929 H.E. Ives and colleagues at Bell Telephone Laboratories transmitted 50-line colour television images between New York City and Washington, D.C.; this also was a mechanical method, but one that sent simultaneously the three primary colour signals over three separate circuits. In the same year Frank Gray, also of Bell Laboratories, applied for a patent that described a method of transmitting two or more signals over a single channel; this introduced important new principles that were to be the foundation of modern (compatible) colour television as it was developed about 20 years later.

It soon became apparent that two basic approaches to colour television were possible. One was the frame-by-frame sequential transmission of signals corresponding to each of the three primary colours. This method was in most respects relatively simple to achieve but involved an increased rate of scanning in order to avoid colour flicker, with resulting transmission difficulties from both the higher bandwidth and the inability to use existing black-and-white receivers to reproduce any pictures originated in colour. This kind of system is therefore usually described as noncompatible. The alternative approach—practically much more difficult, even daunting at first—recognized the advantage to be derived from a colour transmission system in which signals representing the three primary colours were transmitted simultaneously and which could also be compatible with existing black-and-white transmissions. Such a system would mean that any pictures originated in colour could still be receivable (in black and white) on a black-and-white receiver.

Compatible and noncompatible colour television

In 1938 Georges Valensi of France pioneered the path to compatible colour television when he patented a method that enabled

> the output from a single transmitter to be received not only by television receivers provided with the necessary equipment but also by the ordinary type of receiver which is more numerous and less expensive and which reproduces the pictures in black and white.

Although Valensi's proposals have not been precisely adopted in practice, they were influential in later approaches to the problem of compatibility.

Experimental work on colour (on high-definition standards) was taken up in both Great Britain and the United States during the late 1930s. Similar methods were explored by Baird in Great Britain and Peter Goldmark of the Columbia Broadcasting System (CBS) in the United States, both of whom demonstrated sequential systems using rotating colour filters on the cameras and the receivers. The CBS method was used for some experimental broadcasts before World War II, and these were resumed in 1951 as a service authorized by the Federal Communications Commission (FCC). Public interest was small, however, and after only a few months the broadcasts were abandoned.

Serious attention was then given by the National Television Systems Committee (NTSC) in the United States to the development of a fully compatible simultaneous system. The committee's work led in 1953 to a system capable of operating within the current black-and-white standards and was accepted by the FCC. The essentials of this system, known as the NTSC system, have formed the basis of colour systems throughout the world. The basic principle is the combination of two image transmissions: one carrying information about the brightness, including the finest details of the televised scene, to which black-and-white receivers respond; the other, of coarser structure, carrying the colour information. This second component has no appreciable effect on black-and-white sets, while colour receivers use a combination of the two image transmissions. Since the ability of the human eye to perceive detail is most acute when viewing white light, the brightness component of the colour images carries the impression of fine detail, and the superimposed coarse colour information does not substantially alter the sharpness of the resulting colour picture.

In the United States, public broadcasting using the NTSC system began in 1954. The same system was adopted by Japan, where it came into service in 1960. Other countries favoured modifications of the NTSC system. One such was devised by W. Bruch of the German Telefunken Company; known as PAL (phase alternation line), this utilized a small but subtle variant of the NTSC method. The other, SECAM (*système électronique couleur avec mémoire*), differing rather more radically, had been put forward earlier by Henri de France in Paris. Both alternatives aimed at reducing the sensitivity of the colour system to certain forms of distortion encountered in transmission and broadcasting, and had special application to European conditions. Countries were divided in their preferences: in 1967 Great Britain and the Federal Republic of Germany began colour broadcasting using the PAL system, while in the same year France and the Soviet Union also introduced colour, adopting the SECAM system. One or the other of these three systems was adopted by all countries in the 1970s.

The SECAM system

Colour television is technically much more complex than black and white. In the early 1970s most cameras embodied either three or four pickup tubes, while in the receiver a shadow-mask, three-gun, cathode-ray tube, first demonstrated in 1950, formed a successful and widely adopted type of display device for domestic purposes. Widespread purchase of colour receivers began in the United States in 1964. (T.H.B.)

PRINCIPLES OF PICTURE TRANSMISSION AND RECEPTION

Basic factors. The quality and quantity of television service are limited fundamentally by the rate at which it is feasible to transmit the picture information over the television channel. In modern practice the televised image must be capable of being dissected, within a few hundredths of a second, into more than 100,000 picture elements. This implies that the electrical impulses corresponding to the picture elements must pass through the channel at a rate as high as several million per second. Moreover, since the picture content may vary, from frame to frame, from simple close-up shots having little fine detail to comprehensive distant scenes in which the limiting detail of the system comes into play, the actual rate of transmitting the picture information varies from time to time, from a few impulses per second to several million per second. The television channel must be capable, therefore, of handling information over a continuous band of frequencies several million cycles wide. This is testimony to the extraordinary comprehension of the human sense of sight. Hearing is comparatively crude. The ear is satisfied by sound produced by impulses that can be carried over a channel only 10,000 cycles wide.

In the United States, the television channel occupies a width of six megacycles (6,000,000 hertz, or cycles per second) in the radio spectrum. This is 600 times as wide as the channel used by each standard sound broadcasting station. In fact, each television station uses nearly six times as much spectrum space as all the commercial amplitude-modulation (AM) sound broadcasting channels combined. Since each television station must occupy so much spectrum space, few channels are available in a given locality. Moreover, the quantity of service is in conflict with the quality of reproduction. If the detail of the television image is to be increased, other parameters of the transmission being unchanged, the channel width must be increased proportionately, and this decreases the number of channels that can be accommodated in the spectrum. This fundamental conflict between quality of transmission and number of available channels dictates that the quality of reproduction shall just satisfy the typical viewer under normal viewing conditions. Any excess of performance beyond this ultimately would result in a restriction of program choice.

Flicker. The first requirement to be met in image analysis is that the reproduced picture shall not flicker, since flicker induces severe visual fatigue. Flicker becomes more evident as the brightness of the picture increases. If flicker is to be unobjectionable at brightness suitable for home viewing during daylight as well as evening hours (25 to 100 footlamberts), the successive illuminations of the picture screen should occur no fewer than 50 times per second. This is approximately twice the rate of picture repetition needed for smooth reproduction of motion. To avoid flicker, therefore, twice as much channel space is needed as would suffice to depict motion. The same disparity occurs in motion-picture practice, in which satisfactory performance with respect to flicker requires twice as much film as is necessary for smooth simulation of motion.

rojecting
ach pic-
ure twice

A way around this difficulty has been found, in motion pictures as well as in television, by projecting each picture twice. In motion pictures, the projector interposes a shutter briefly between film and lens while a single frame of the film is being projected. In television, each image is analyzed and synthesized in two sets of spaced lines, one of which fits successively within the spaces of the other. Thus the picture area is illuminated twice during each complete picture transmission, although each line in the image is present only once during that time. This technique is feasible because the eye is comparatively insensitive to flicker when the variation of light is confined to a small part of the field of view. Hence flicker of the individual lines is not evident. If the eye did not have this fortunate property, a television channel would have to occupy about twice as much spectrum space as it now does.

It is thus possible to avoid flicker and simulate rapid motion by a picture rate of about 25 per second, with two screen illuminations per picture. The precise value of the picture-repetition rate used in a given region has been chosen by reference to the electric power frequency that predominates in that region. In Europe, where 50-hertz power is the rule, the television picture rate is 25 per second (50 screen illuminations per second). In North America the picture rate is 30 per second (60 screen illuminations per second) to match the 60-hertz power that predominates there.

The higher picture-transmission rate of North America allows the pictures there to be about five times as bright as those in Europe for the same susceptibility to flicker, but this advantage is offset by a 20 percent reduction in picture detail for equal utilization of the channel.

Image structure. The second aspect of performance to be met in a television system is the detailed structure of the image. A printed halftone engraving may possess several million halftone dots per square foot of area. Such reproductions are intended for minute inspection, and the dot structure must not be apparent to the unaided eye even at close range. Such fine detail would be a costly waste in television, since the television picture is viewed at comparatively long range (five to 15 feet, or 1.5 to 4.5 metres, in the typical home setting) and the picture area is not greater than about two square feet (1,800 square centimetres). Under these conditions, a picture structure of about 200,000 halftone elements is a suitable compromise. This detail is about equal to that provided by 16-millimetre home motion pictures and substantially exceeds that of the eight-millimetre film widely used for amateur cinematography.

Picture shape. The third item to be selected in image analysis is the shape of the picture. This has been standardized universally as a rectangle the width of which is one-third longer than its height. This 4:3 ratio (aspect ratio) was originally chosen to match the dimensions of standard 35-millimetre motion-picture film (prior to the advent of the wide-screen cinema) in the interest of televising film without waste of frame area. The width of the screen rectangle is greater than its height, as in the proscenium of a theatre, to accommodate the horizontal motion that predominates in virtually all televised events.

Scanning. The fourth determination in image analysis is the path over which the image structure is explored at the camera and reconstituted on the receiver screen. In the original Nipkow disk, this path was a series of arcs of circles (Figure 1). In modern television, the pattern is a series of parallel straight lines, each progressing from left to right, the lines following in sequence from top to bottom of the picture frame. The exploration of the image structure proceeds at a constant speed along each line, since this provides uniform loading of the transmission channel under the demands of a given structural detail, no matter where in the frame the detail lies. The line-by-line, left-to-right, top-to-bottom dissection and reconstitution of television images is known as scanning, from its similarity to the progression of the line of vision in reading a page of printed matter. The agent that disassembles the light values along each line is called the scanning spot, and the path it follows is the scanning pattern, or raster.

The geometry of the modern scanning pattern is shown in Figure 2. It consists of two sets of lines. One set, marked A, is scanned first, and the lines are so laid down that an equal empty space is maintained between lines. The second set, marked B, is laid down after the first and is so positioned that its lines fall precisely in the empty spaces of set A. The area of the image is thus scanned twice, but each point in the area is passed over once. This is known as interlaced scanning, and it is used in all the television broadcast services of the world. Each set of alternate lines is known as a scanning field; the two sets together, comprising the whole scanning pattern, are known as a scanning frame. The repetition rate of field scanning is standardized in accordance with the frequency of electric power, as noted above, at either 50 or 60 fields per second. Corresponding rates of frame scanning are 25 and 30 frames per second. The localities for various scanning standards are listed in Table 6.

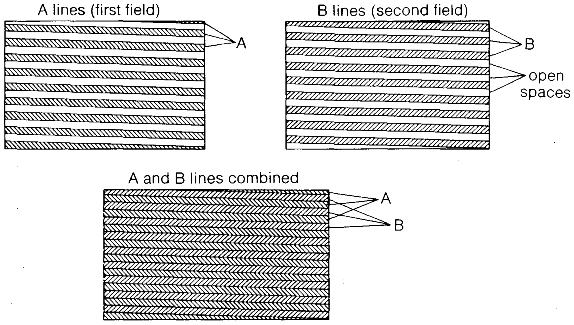

Figure 2: Interlaced scanning, combining two separate scanning cycles (fields) to form a complete picture.

Interlaced
scanning

The total number of lines in the scanning pattern has been set to provide a maximum pictorial detail of the order of 200,000 picture elements. Since the frame area is four units wide by three units high, this figure implies a pattern of about 520 picture elements in its width (along each line) and 390 elements in its height (across the lines). This latter figure would imply a scanning pattern of about 400 lines, were it not for the fact that many of the picture details, falling in random positions on the scanning pattern, lie partly on two lines and hence require two lines for accurate reproduction. Scanning patterns are designed, therefore, to possess about 40 percent more lines than the number of picture elements to be reproduced on the vertical direction. Actual values in use in television broadcasting in various regions are 405 lines, 525 lines, 625 lines, and 819 lines per frame, as listed in Table 6. These values have been chosen to suit the frequency band of the channel actually assigned in the respective geographical regions.

The scanning spot is made to follow the paths shown in Figure 2 by imparting two repetitive motions to the spot simultaneously (Figure 3). One is a horizontally directed

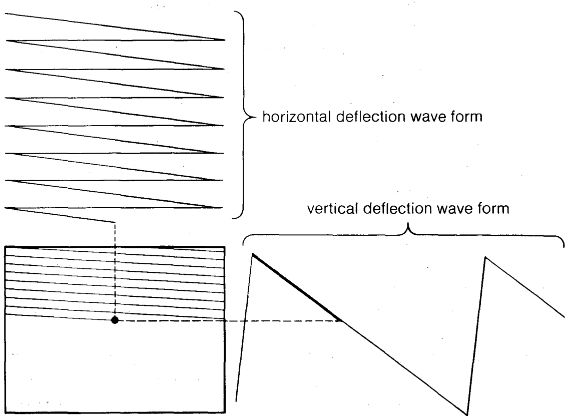

Figure 3: Scanning pattern and wave forms for horizontal and vertical deflection in sequential parallel-line scanning.

back-and-forth motion. The spot is moved at constant speed from left to right and then returned as rapidly as possible, while extinguished and inactive, from right to left.

Simultaneously, a vertical motion is imparted to the spot, moving it at a comparatively slow rate from the top to the bottom of the frame. This motion spreads out the more rapid left-to-right scans, forming the first field scan of alternate lines and empty spaces. When the bottom of the frame is reached, the spot moves vertically upward as rapidly as possible, while extinguished and inactive. The next top-to-bottom motion then spreads out the horizontal line scans so that they fall in the empty spaces of the previously scanned field. Precise interlacing of the successive field scans is facilitated if the total number of lines in the frame is an odd number. All the numbers of lines used in ordinary transmissions (see Table 6) were chosen for this reason.

The return of the scanning spot from right to left and from bottom to top of the frame, during which it is inactive, consumes time that cannot be devoted to transmitting picture information. This time is then used to transmit synchronizing control signals that keep the scanning process at the receiver in step with that at the transmitter. The amount of time lost during retracing of the spot proportionately reduces the actual number of picture elements that can be reproduced. In the 525-line scanning pattern used in the United States, about 15 percent of each line is lost in the return motion, and about 35 out of the 525 lines are blanked out while the spot returns from bottom to top of two successive fields. The scanning area that is actually in use for reproduction of the picture contains a maximum of about 435 picture elements along each line and has 490 active lines, capable of reproducing 350 picture elements in the vertical direction. The frame can accommodate at maximum, therefore, about 350 × 435, or 152,000, picture elements. The detail-transmission capacity of the scanning patterns of several other systems is listed in Table 6.

Time needed for scanning

The time taken by the scanning spot to move over the active portion of each scanning line is of the order of 0.00005 second (50 microseconds). In the United States system, 525 lines are transmitted in one-thirtieth of a second, which is equivalent to about 64 microseconds per line. Up to 15 percent of this time is consumed in the horizontal retrace motion of the spot, leaving 54 microseconds (54×10^{-6} seconds) for active reproduction of as many as 435 picture elements in each line. This represents a maximum rate of $435 \div 54 \times 10^{-6} \cong 8{,}000{,}000$ picture elements per second.

Table 6: Television Systems of the World

region or country	number of lines per frame	number of pictures per second	maximum detail (picture elements per frame)	available picture band-width (MHz)	channel band-width (MHz)
United Kingdom	405–625	25	130,000–210,000	3–6	5–8
North America, South America, Japan	525	30	130,000	4	6
Europe, Australia, Africa, Eurasia	625	25	210,000	6	8
France and French dependencies	625–819	25	210,000–440,000	6–10.4	8–14

Two picture elements can be approximately represented by one cycle of the transmission signal wave. The signal must therefore be capable of carrying components as high as 4,000,000 hertz (four megahertz, MHz). The United States six-megahertz television channel provides a sufficient band of frequencies for this (plus an additional two megahertz for transmission of the sound program, to protect against interference, and to meet the requirements of vestigial side-band transmission).

The relationship between the ideal and actual scanning patterns is shown in Figure 4. The part shown as darker is lost as the spot retraces. The remaining area of the pattern is actively employed in analyzing and synthesizing the picture information and is adjusted to have the four by three dimensions of the standard aspect ratio. In practice, some of the active area may also be hidden behind the decorative mask that surrounds the picture tube of the receiver, as shown by the dashed line.

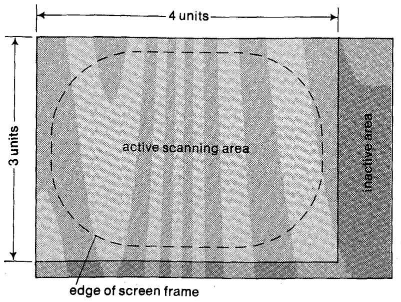

Figure 4: Active and inactive portions of scanning pattern.

The television picture signal. The signal wave form (the sequence of electrical impulses that makes up a television picture signal) embodies all the picture information to be transmitted from camera to receiver screen as well as the synchronizing information required to keep the receiver and transmitter scanning operations in exact step with each other. The television system, therefore, must deliver the wave form to each receiver as accurately and as free from blemishes as possible. Unfortunately, almost every item of equipment in the system (amplifiers, cables, transmitter, the transmitting antenna, the radio wave in space, the receiving antenna, and the receiver circuits) conspires to distort the wave form or permits it to be contaminated by "noise" or interference.

Distortion and interference

Among the possible distortions in the signal producing the picture are: (1) failure to maintain the rapidity with which the wave form rises or falls as the scanning spot crosses a sharp boundary between light and dark areas of the image, producing a loss of detail or smear in the reproduced image; (2) the introduction of overshoots, which cause excessively harsh outlines; and (3) failure to maintain the average value of the wave form over extended periods, which causes the image as a whole to be too bright or too dark.

Throughout the system, amplifiers must be used to keep the television signal strong relative to the random electrical currents (noise) that are everywhere present. These random currents, caused by thermally induced motions of electrons in the circuits, cause interference having a speckled appearance, known as "snow." Pictures received from distant stations are subject to this form of interference, since the radio wave is then so weak that it cannot override the random currents in the receiving antenna. Interference of a striated type may be caused by the signals of stations other than that to which the receiver is tuned. Care in the design of the receiver tuner and amplifier circuits is necessary to minimize such interference, and channels must be allocated to neighbouring communities at sufficient geographical separations and frequency intervals to protect the local service.

The television signal wave form can be conveyed within transmitting and receiving apparatus, and over short distances, by wires and cables. For transmission over greater distances, however, as in network connections and over the air, the television signal wave form must be imposed on a high-frequency carrier wave. The broadcast transmitter, for example, is actually a generator of very-high-frequency or ultrahigh-frequency radio waves. When such a current (known as the carrier current) is passed through the transmitting antenna, it produces a radio wave of the same frequency that travels through space and induces a weaker, but otherwise identical, current in the receiving antenna.

At the transmitter the picture signal wave form changes the strength, or amplitude, of the high-frequency carrier current by a process known as amplitude modulation. The alternations of the carrier current are thus constrained to take on a succession of amplitudes that match the shape

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of the signal wave form. The functions of the television system that generate, transmit, and utilize the television signal wave form are shown in Figure 5. The scene to be televised is focussed by a lens on a camera tube, located within the camera. A scanning spot within the camera tube is passed over the image in the scanning pattern that is shown in Figure 3, its motion being controlled by the synchronization generator. The latter equipment also renders the scanning spot inactive during its retrace motions and generates the necessary synchronization signals. The camera produces the signal wave form; the synchronizing signals are then added, and the complete signal wave form is established. The signal is then amplified as required, by amplifiers that are capable of operating over the wide band of frequencies (from a few hertz to as many as 4,000,000 or 5,000,000 hertz) that are necessary to preserve the precise shape of the wave form. In travelling along a transcontinental network, the signal may pass through more than 100 such amplifiers in tandem before it finally reaches the transmitter. There the wave form is amplified to a high power level. At the same time, the carrier current is generated by other circuits of the transmitter. The signal wave form and carrier current are then combined in the modulator, which imposes the wave form on the carrier. The modulated carrier current may then be further amplified (typically to 10,000 watts or more) and finally passed through the transmitter antenna, which radiates the television radio wave. A similar but enormously weaker current is induced in any receiving antenna that may be located within the service area. Reception of excellent quality is possible when the power picked up by the receiving antenna is as low as 0.00000001 (10^{-8}) watt. The power picked up from a distant station may be as low as 10^{-11} watt; but even this low power is capable of producing reception of acceptable quality.

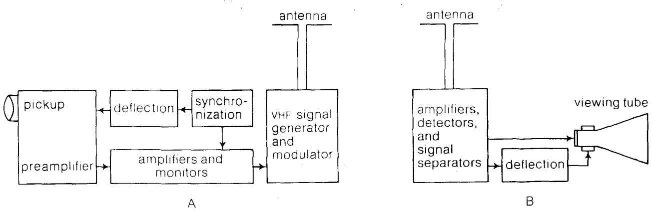

Figure 5: Essential elements of (A) transmitter, and (B) receiver.

The transmitter antenna must be placed to stand as high and in as exposed a location as possible, since the radio waves tend to be intercepted by solid objects that stand in their path, including the Earth's surface at the horizon. Reception beyond the horizon is possible, but the signal at such distances becomes rapidly weaker as it passes to the limit of the service area.

In the receiver, the current from the antenna is passed through a tuner that separates it from other currents of different carrier frequencies that may be present from other stations. The desired current is then amplified in several successive amplifier stages, while still in the amplitude-modulated form. Passage through a detector follows. The picture signal wave form is thereby recovered and is passed through still another amplifier before passing to the picture tube. The picture screen displays a scanning spot the brightness of which is controlled by the picture signal. Simultaneously, the signal is applied to synchronizing circuits, which abstract from it the synchronizing signals. These are used to control the motion of the scanning spot so that it traces out the scanning pattern in precise synchronism with the motion of the scanning spot in the camera.

The operations governing transmission and reception of the sound program accompanying the television's picture signal are accomplished by the use of apparatus similar to that used for frequency-modulated (FM) radio broadcasting. The carrier frequency for this sound channel is usually spaced 4.5 megahertz from the picture carrier and is separated from the picture carrier in the receiver by appropriate circuitry.

The amount of amplification conferred on the picture and sound currents by a typical television receiver is extremely large. When tuned to a station at a distance of 50 miles (about 80 kilometres), the power picked up by the antenna is typically 10^{-11} watt, as noted above, whereas the signals fed to picture tube and loudspeaker, after amplification, are of the order of one watt; that is, the receiver produces a faithful amplification of the order of 100,000,000,000 times.

Common
elements of
television
camera
tubes

Television camera tubes. The television camera tube is an electronic device that converts an optical image into a sequence of electrical impulses; *i.e.,* it produces the television picture signal. Three types of television camera tube are in wide use: the image orthicon, the orthicon, and the Vidicon. These devices differ to some extent in detail, but all have in common the following elements: (1) a photosensitive surface (on which the scene is focussed) that converts each light value into a corresponding value of electric charge and forms thereby an electrical image; (2) a means of storing the electrical image—that is, a way of causing the charge to accumulate during the interval of several hundredths of a second between successive scannings of each line in the image; (3) an electron beam, formed in an electron gun and deflected over the electrical image, following the scanning pattern of Figure 2; and (4) a mechanism for producing an electrical current or voltage that is proportional at every instant to the electric charge accumulated at the point passed over by the electron beam at that instant.

The television camera tube is also capable of extinguishing the scanning beam during its retrace motions, so that no television signal is then generated. A television pickup device that does not store the electrical image is called the flying spot scanner.

Iconoscope. The iconoscope is a type of camera tube that is no longer used, since it is not as sensitive as the image orthicon and its images are subject to uneven shading and flare. But it is well suited to introduce the concepts of electron image storage and scanning in simple form. The iconoscope is housed in a dipper-shaped, vacuum-tight glass envelope, as shown in Figure 6. Within the wide end is a flat sheet of mica. A uniform metallic coating, called the signal plate, is placed on the rear surface of this sheet, away from the image. The front surface of the mica is covered with a mosaic composed of many hundreds of thousands of tiny globules of silver.

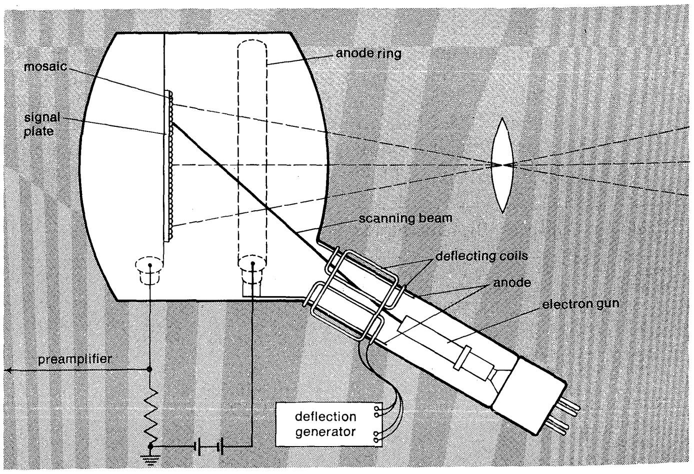

Figure 6: Iconoscope camera tube.

During the manufacture of the tube the mosaic is treated with cesium vapour and oxygen, so that each globule has a surface of the oxides of silver and cesium. This combination of elements provides a surface from which electrons are readily liberated, by the photoelectric effect, when light falls on it. Since the globules are insulated from each other and from the signal plate by the mica, the loss of electrons under illumination causes the globules to assume and hold a positive charge, the charge on each globule being proportional to the strength of the illumination falling on it and to the time it has been illuminated.

When an optical image is focussed on the treated mosaic, the whole surface assumes a distribution of positive charge that corresponds to the distribution of light in the image. The amount of charge at each point on the surface steadily increases, if the optical image is maintained, until the scanning spot passes over the globule of silver at that point.

Iconoscope scanning spot

The scanning spot of the iconoscope is formed by a narrow beam of electrons, shot out of an electron gun in the side arm of the tube. On its way to the mosaic, this beam passes within two sets of electromagnet coils. Currents like those of Figure 3 are passed through these coils, causing the beam to be deflected horizontally at a rapid rate and vertically at a relatively slower rate. The extent of the horizontal motion is adjusted from top to bottom of the mosaic, so that the pattern traced out by the electron beam on the mosaic is a rectangular pattern like that shown in Figure 2.

As each globule is passed over by the beam, it undergoes a sudden change in electrical potential, the amount of the change being proportional to the light falling on it. The change in potential of the globule is transferred through the mica support to the signal plate behind it, the globule and plate forming in effect the plates of an electrical capacitor. Thus, as the beam passes in succession over the globules lying along a given scanning line, the signal plate assumes a succession of voltages (the picture signal) that match the corresponding succession of light values along that line. The signal plate is connected to an amplifier, external to the iconoscope, that increases the strength of the picture signal.

The phenomenon of charge storage, by which the magnitude of the electrical image is continually increased between successive scannings of each line, is of the utmost significance in television technology. The spirally apertured rotating Nipkow disk (and other non-storage television pickup devices) employs only the light that is present at a given point in the image at the instant the scanning spot passes over that point. Since in modern television the area of the scanning spot is only about one two-hundred-thousandth of the area of the scanning pattern, only this small fraction of the light of the image can be used. But when the image charge is stored in increasing amount for the full interval between successive scannings of a given point, the accumulated charge is then theoretically increased by about 200,000 times the single charge that can be accumulated during the time the beam moves through its own width.

Orthicon. The orthicon camera tube was the first successor to the iconoscope. This tube (Figure 7) is housed in a cylindrical glass structure, on one end of which the optical image is focussed on a mosaic. This mosaic is similar to that of the iconoscope but is made up of a precise array of squares of photosensitive material, formed by evaporation of the material onto a transparent support. A transparent metal coating on the reverse side serves as the signal plate.

By courtesy of RCA

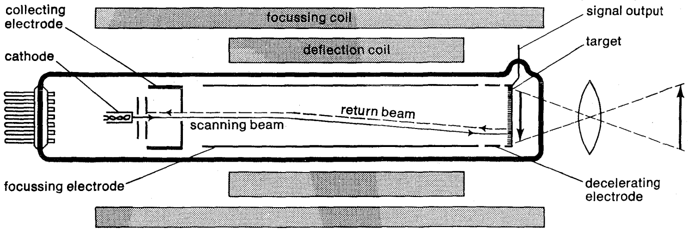

Figure 7: Orthicon camera tube.

The optical image is focussed on the signal plate, passing through it and the support to the mosaic, where it creates and stores a positive electrical charge image, in the manner of the iconoscope. The electron beam is directed toward the mosaic and deflected by electromagnet coils surrounding the tube, so that it passes over the mosaic in the standard interlaced scanning pattern. The electrons are accelerated by a low voltage, so that they strike the mosaic with a low velocity and do not liberate any substantial number of secondary electrons. Instead, the negative electrons in the beam land on the mosaic squares, neutralizing the stored positive charge and ultimately building up a negative charge. The electrons in the beam are then repelled and return to the opposite end of the tube, where they are removed by a positive collecting electrode that surrounds the electron gun. The change in electrical potential, undergone by each mosaic element as its charge is neutralized, is transferred by capacitive action, as in the iconoscope, to the signal plate. The latter thereby undergoes a succession of voltage changes that constitute the picture signal and that are passed to amplifier stages external to the tube.

Sensitivity of the image orthicon

Image orthicon. The image orthicon is the most highly developed of the television camera tubes and is perhaps the most remarkable electronic device in existence. In its most refined form, its sensitivity to light is phenomenally great. It can respond to light levels far below those capable of exposing motion-picture film or affecting the eye itself. Less sensitive image orthicon tubes are used in broadcasting practice; they serve to produce satisfactory image quality under any light levels likely to be encountered in studio and outdoor scenes.

The image orthicon, shown in Figure 8, is similar to the orthicon but includes an additional electrical-imaging process and contains a high-gain amplifier based on the phenomenon of electron multiplication. It is housed in a cylindrical glass envelope having an enlarged section on one end, which is closed by an optically flat glass plate on the inside of which is deposited a continuous photosensitive coating (photocathode). The optical image is focussed through the glass support onto the coating, where it liberates electrons by photoelectric emission, the emission at each point matching the amount of light in the image at that point.

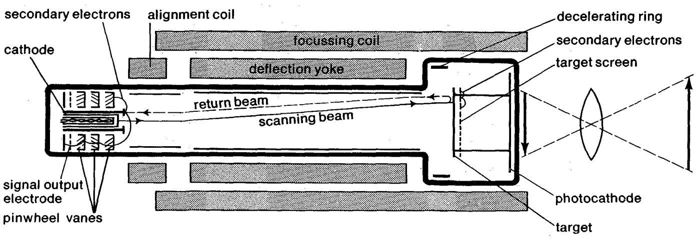

Figure 8: Image orthicon camera tube.

The liberated electrons stream away from the photosensitive surface of the glass plate, passing through the enlarged cylindrical portion of the tube, until they encounter the target electrode, a piece of very thin glass of uniform thickness. The streams of electrons induce secondary emission of electrons from the surface of the glass, and these secondary emissions are collected by a fine-mesh screen that lies parallel and close to the glass target. The streams of image electrons then pass through the screen on their way to the target.

The changes in electrical potential induced by secondary emission are transferred to the opposite face of the target, where they are scanned by the electron beam. Because of the characteristics of the glass target, the fine structure of the electrical image is preserved throughout the scanning interval. Moreover, as the streams of electrons from the photocathode continue to fall on the target glass, the intensity of the electrical image continually increases; that is, storage of charge occurs.

Both the photocathode and the glass target are continuous surfaces and in themselves they contribute no structural limit to the detail of the image. The fine-mesh screen that collects the secondary electrons from the glass target could, however, impose a limit on detail, and consequently the holes in the screen must be substantially smaller than the scanning spot. Actually, the screen possesses a total of nearly 1,000,000 holes in an area of 1.5 square inches (about 10 square centimetres), compared with the roughly 200,000 picture elements into which the optical image is dissected.

The stored charge image on the reverse side of the target is scanned by an electron beam that, like that of the orthicon, is accelerated at lower voltage and is actually slowed down before hitting the target, so that the scanning electrons have a very low velocity when they strike the target. Consequently, the secondary emission caused by scanning is negligible.

In the standard interlaced scanning pattern the electron beam is magnetically deflected over the electrical image by coils external to the tube. When the beam electrons hit the target, they neutralize the stored positive charge, as in the orthicon, and when equilibrium is reached, further electrons may not land on that spot. Thereafter, the electrons in the beam are turned around at the target and return to the vicinity of the electron gun at the other end of the tube. At this juncture the action differs from that of the image orthicon. Instead of taking the picture signal from a signal plate at the glass target, the return beam of electrons is used. The number of electrons returning from each point in the scanning pattern is conditioned by the prior positive charge neutralization occurring at that point. Consequently, the return beam constitutes a picture signal current.

The return beam

When the return beam reaches the opposite end of the tube, it is passed through an amplifier structure that contributes greatly to the sensitivity of the tube. This structure, known as an electron multiplier, is composed of several metal pinwheels. The surfaces of the pinwheel vanes are treated to emit secondary electrons copiously when the return-beam electrons hit them; more electrons are emitted than strike the surface. The electron multiplier's pinwheel shape causes the secondary electrons liberated at the first vane to be deflected to the next vane in the same structure, where still further emission of secondary electrons occurs. The magnified stream then passes to the next pinwheel structure to the rear, where the process repeats and the electron stream is cumulatively multiplied. The amplified current at the final stage of the electron multiplier is passed out of the tube to a conventional vacuum-tube or transistor amplifier.

Vidicon. The Vidicon camera tube, first used in the early 1950s, was the first camera device to employ the phenomenon of photoconductivity. In its early form it was sluggish (that is, objects in motion were reproduced with noticeable smear), and hence for a time it was limited to industrial applications. Later versions were free of this defect, and the Vidicon now enjoys wide use in broadcasting, in both monochrome and colour service. The tube elements (Figure 9) are contained in a cylindrical glass envelope, substantially smaller than the image orthicon and hence more adaptable to portable cameras. At one end of the envelope, a transparent metallic coating serves as a signal plate. Deposited directly on the signal plate is a photoconductive material (*e.g.*, a compound of selenium or lead) the electrical resistance of which is high in the dark but becomes progressively less as the amount of light increases.

By courtesy of RCA

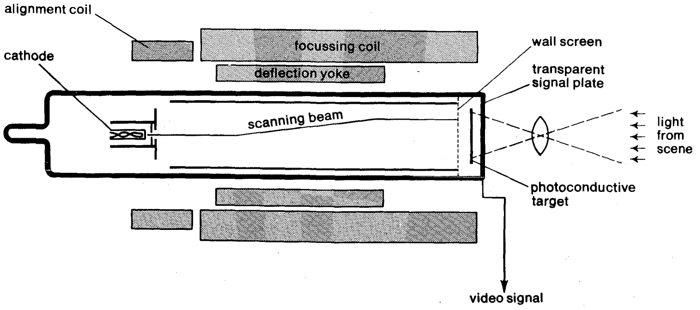

Figure 9: Vidicon camera tube.

The optical image of the Vidicon

The optical image is focussed on the end of the tube and passes through the signal plate to the photoconductive layer, where the light induces a pattern of varying conductivity that matches the distribution of brightness in the optical image. The conduction paths through the layer allow positive charge from the signal plate (which is maintained at a positive voltage) to pass through the layer, and this current continues to flow during the interval between scans. Charge storage thus occurs, and an electrical charge image is built up on the rear surface of the photoconductor.

An electron beam, deflected in the standard interlaced scanning pattern, scans the rear surface of the photoconducting layer. The beam electrons neutralize the positive charge on each point in the electrical image, and the resulting change in potential is transferred by capacitive action to the signal plate, from which the television signal is derived.

The electron beam scanning spot size imposes the only limit on the image detail. It is possible, therefore, to derive an image of broadcast quality (200,000 or more picture elements) from a photosensitive area no larger than 0.2 square inch (1.3 square centimetres). This permits the use of small and comparatively inexpensive lenses, with correspondingly large depth of focus for a given lens opening. The internal structure of the tube is very simple, and this, with its small size, makes it adaptable to a wide range of camera arrangements in broadcasting, industrial, and military applications.

Flying spot scanner. Another form of television pickup device, used to transmit images from film transparencies, either still or motion pictures, is the flying spot scanner shown in Figure 10. The light source is a cathode-ray tube in which a beam of electrons, deflected in the standard scanning pattern, impinges on a fluorescent phosphor surface. The beam produces thereby a spot of light that moves in the scanning pattern on the face of the tube. The light from this spot is focussed optically on the surface of the photographic film transparency to be televised. As the image of the spot moves, it traces out a scanning line across the film, and the amount of light emerging from the other side of the film at each point is determined by the degree of transparency of the film at that point. The emerging light is collected and caused to enter a photoelectric cell, which produces a current proportional to the light entering it.

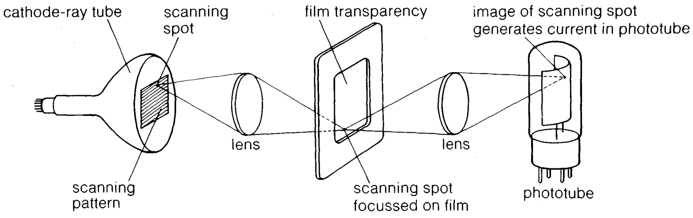

Figure 10: Flying spot camera system.

This current thus takes on a succession of values proportional to the successive values of film density along each line in the scanning pattern; that is, it is the picture signal current. No storage action occurs, so the light from the cathode-ray tube must be very intense and the optical design very efficient to secure noise-free reproduction. The flying spot system may be used with motion-picture film if an optical immobilizer is used.

Video amplification. The picture signal wave form produced by the television camera is initially very weak, of the order of a few hundredths of a volt, and is therefore vulnerable to the random currents (noise) present in the succeeding transmission circuits. Immediately upon leaving the camera, therefore, the picture signal must be strengthened, and this process must be repeated at intervals throughout the transmission to the picture tube at the receiver. The strengthening process, known as video amplification, takes place in electron tubes or transistors (see ELECTRONICS).

Video amplifiers, of both the tube and transistor varieties, can be designed to cover a remarkably comprehensive range, amplifying a picture signal consisting of any combination of frequency components lying from 10 hertz to as high as 10,000,000 hertz. This implies not merely that the amplitudes of all such components are uniformly increased but that their relative times of occurrence are strictly preserved. Only if these requirements are met can the fine structure, as well as the gross outline, of the picture signal wave form be preserved from the camera to the receiver picture tube.

Side bands

The television channel. When the band of frequencies in the picture signal is imposed on the high-frequency broadcast carrier current in the modulator of the transmitter, two bands of frequencies are produced, above and below the carrier frequency as described earlier. The side bands are identical in frequency content; that is, both carry the complete picture signal information. One of the side bands is superfluous and, if transmitted, would wastefully consume space in the broadcast spectrum. Therefore, the major portion of one of the side bands is removed, and the other is transmitted in full. Complete removal of the superfluous side band is possible, but this would complicate receiver design. Hence, a vestige of the unwanted side band is retained to serve the overall economy of the system. This technique is known as vestigial side-band transmission. It is now universally employed in all of the television broadcasting systems of the world.

The television channel for a black-and-white picture (monochrome) contains the picture carrier frequency, one complete picture side band, and a vestigial portion of the other picture side band. In addition, the carrier for the sound transmission and its side bands is included within the channel. Since the band of frequencies needed to convey the sound is a very much narrower band than that needed for the picture, it is feasible to include both sound-carrier side bands. Frequency modulation is employed to transmit the sound information. To avoid mutual interference between sound and picture, the picture and sound side bands must not overlap. Moreover, some space must be allowed at the edge of the channel to avoid interference with the transmissions of stations occupying the adjacent channels.

These requirements are met in the television channel shown in Figure 11. Each contains the following bands:

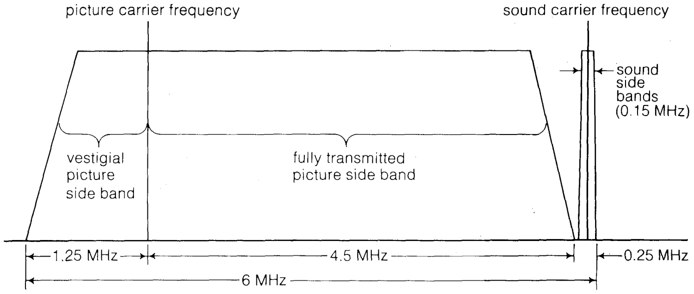

Figure 11: Allocation of television channel for monochrome broadcasting in the United States.

4.5 megahertz for the fully transmitted picture side band, 1.25 megahertz for the vestige of the other picture side band, 0.15 megahertz for the sound carrier and its two side bands, and the remaining 0.1 megahertz to guard against overlap between picture and sound and between channels. In compatible colour transmissions an additional colour-information signal, shown as the chrominance subcarrier, is included within the fully transmitted picture side band.

The television channels of the United States are assigned 6 megahertz each in the following segments of the spectrum: VHF (very-high-frequency) channels 2, 3, and 4, 54–72 megahertz; 5 and 6, 76–88 megahertz; 7 through 13, 174–216 megahertz; and the UHF (ultrahigh-frequency) channels, 14 through 83, 470–890 megahertz. These channels are allocated to communities according to a master plan established and administered by the Federal Communications Commission. No more than seven VHF channels are provided in any one area (this number is allocated only to New York and Los Angeles). Many smaller cities must be content with one or two channels.

In the major cities of Europe, fewer channels (typically two to four per city) are provided, because the higher population density and closer spacing of cities precludes more assignments within the available spectrum and because fewer program services are offered. In most countries the programs are produced solely by government agencies, in contrast to the commercial sources of the United States. Program services competitive to the government outlets exist in Great Britain, Canada, Japan, and in some countries of South America. Tokyo, for example, has seven channels.

The television transmitter. The task of the television transmitter is to produce the picture and sound carriers, and their side bands, in conformity with the arrangement of the channel assigned to it. The electronic equipment that performs these functions is shown in Figure 12. The central element is the synchronizing generator, which controls the timing of the scanning of the camera and produces the synchronizing impulses. It applies three control functions to the camera: timing for horizontal (line-by-line) deflection of the scanning beam; timing for vertical deflection (field-by-field); and the blanking impulses that extinguish the scanning spot during its retrace motions. The camera, under the control of the synchronizing generator, produces the picture signal without the synchronizing impulses. The picture signal is passed first through a preamplifier within the camera housing and then through additional amplifiers in the control room of the studio, where it is combined with the synchronizing impulses. At this point, it is customary to inspect the quality of the reproduced image by feeding the signal to a monitor that displays the image on a picture tube. The monitor screen is viewed by the program director and operating personnel so that shortcomings can be corrected.

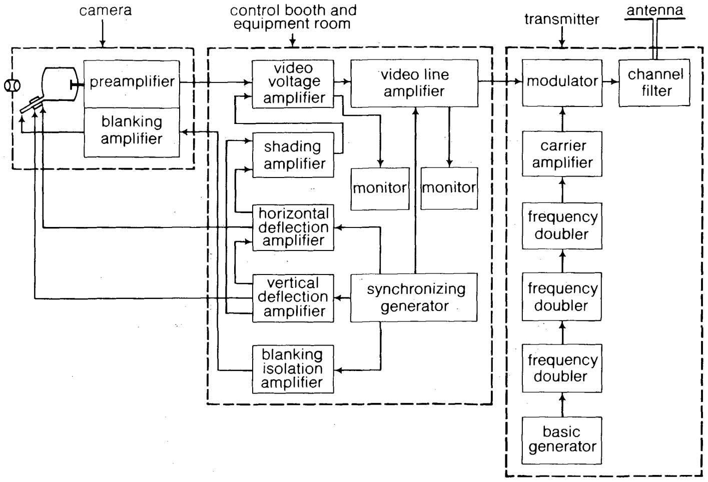

Figure 12: Block diagram of monochrome television transmitter.

The synchronizing generator

The adjusted picture and sound signals are fed to the network that interconnects stations or, in local broadcasts, to cables leading directly to the transmitter. The transmitter is usually located at some distance from the studio to permit the placing of the transmitting antenna in as favourable a location as possible in order to maximize the intended service area.

The transmitter performs two functions: generation of the carrier currents for picture and sound, and imposition of the sound and picture signals on the respective carrier currents. Precautions are taken to assure that the frequencies of the carrier waves have precisely the values assigned to the channel in use. The carrier currents have frequencies accurate to roughly one part in 200,000.

The carrier current and the picture signal are combined in the modulator. The modulated carrier contains two identical side bands; the major portion of one of these (that occupying frequencies below the carrier) is removed by a wave filter, and the remaining part, plus the picture carrier current and the fully transmitted sideband currents, is passed to the transmitting antenna. A similar arrangement produces the sound carrier current and its side bands. These are also fed to the transmitting antenna, which is designed to direct the radiated waves along the surface of the Earth and to minimize radiation toward the sky.

The television receiver. At the receiver the sound and picture carrier waves are picked up by the receiving antenna, producing currents that are identical in form to those flowing in the transmitter antenna, but much weaker. These currents are conducted from the antenna to the receiver by a lead-in transmission line, typically a one-half-inch (12-millimetre) ribbon of polyethylene plastic in which are embedded two parallel copper wires. This form of transmission line is capable of passing the carrier currents to the receiver, without relative discrimination between frequencies, on all the channels to which the receiver may be tuned.

At the antenna terminals of the receiver, the picture and sound signals are at their weakest, so particular care must be taken to control noise at this point. The first circuit in the receiver is a radio-frequency amplifier, particularly designed for low-noise amplification. The channel-switching mechanism (tuner) of the receiver connects one of several individual circuits to this amplifier, each tuned to its respective channel. This amplifier magnifies the voltages of the incoming picture and sound carriers and side bands in the desired channel by about 10 times, and it discriminates by a like amount against the transmissions of stations on other channels.

From the radio-frequency amplifier, the signals are passed to a mixer that transposes the frequencies of the sound and picture carriers to values better suited to the subsequent amplification processes. The transposed frequencies (intermediate frequencies) remain the same, no matter what channel the receiver is tuned to. In typical receivers they are located in the band from 41 to 47 megahertz.

Automatic gain control

Since the tuning of the subsequent amplifiers (intermediate-frequency amplifiers) need not be changed as the channel is switched, they can be adjusted for maximum performance in the 41–47-megahertz range. Two to four stages of such amplification are used in tandem, increasing the voltage of the picture and sound carriers by a maximum of 25 to 35 times per stage, representing an overall maximum amplification of the order of 10,000 times. The amplification of these intermediate-frequency stages is automatically adjusted, by a process known as automatic gain control, in accordance with the strength of the signal, full amplification being accorded to a weak signal, less to a strong signal. After passage through the intermediate amplifiers, the sound and picture carriers and their side bands reach a relatively fixed level of about one volt, whereas the signal levels applied to the antenna terminals may vary, depending on the distance of the station and other factors, from a few millionths to a few tenths of a volt.

From the last intermediate amplifier stage, the carriers and side bands are passed to another circuit, known as the video detector. An averaging circuit or filter then forms, from the detector output, (1) a picture signal, which is a close replica of the picture signal produced by the camera and synchronizing generator in the transmitter, and (2) a frequency-modulated sound signal. Thus far, the picture and sound signals are handled simultaneously in each amplifier and detector circuit. Both may be passed through a video amplifier, if necessary, which increases their voltage by about 10 times.

At this point the picture and sound signals are separated. Another video amplifier passes the picture signal to the picture tube, where it controls the brightness of the scanning spot. Simultaneously, the picture signal wave form is passed to a circuit (the synchronizing signal separator) that selects the blank periods in the wave form containing the superimposed synchronizing signals.

Finally, the sound signal is passed through a sound intermediate amplifier and frequency detector (discriminator, or ratio detector) that converts the frequency modulation back to an audio signal current. This current is passed through one or two additional audio-frequency amplifier stages to the loudspeaker.

Meanwhile, at the synchronizing signal separator the synchronizing signals are passed through circuits that separate the horizontal and vertical synchronizing impulses. They are then passed, respectively, to the horizontal and vertical deflection generators, which produce the currents that flow through the electromagnet deflection coils, causing the scanning spot of the picture tube to be deflected across the viewing screen in the standard scanning pattern.

The synchronizing signals control the timing of these currents, causing the scanning spot of the receiver to move synchronously with that of the camera. Pains are taken in the synchronizing action of the receiver to prevent the scanning motion from being disturbed by interference, since such disturbances would displace the picture elements from their correct positions in the pattern. It is one of the achievements of the modern television system that the scanning pattern is stably reproduced even when the picture signal is so weak and infested with noise that the picture elements themselves are barely recognizable. This fact is of particular significance in colour transmissions.

In addition to the amplifiers, detectors, and deflection generators, the receiver contains two power-converting circuits. One of these (the low-voltage power supply) converts the alternating current from the power line into the direct current needed for the tube or transistor circuits; the other (high-voltage power supply) produces the high voltage, typically 15,000 to 20,000 volts, needed to create the scanning spot in the picture tube.

Viewer's receiver controls

The receiver controls commonly provided for adjustment by the viewer are: (1) the channel switch, which connects the required circuits to the radio-frequency amplifier and superheterodyne mixer to amplify and convert the sound and picture carriers of the desired channel; (2) a fine-tuning control, which precisely adjusts the superheterodyne mixer so that the response of the tuner is exactly centred on the channel in use; (3) a contrast control, which adjusts the voltage level reached by the picture signal in the video amplifiers, producing a picture having more or less contrast (greater or less range between the blacks and whites of the image); (4) a brightness control, which adjusts the average amount of current taken by the picture tube from the high-voltage power supply, thus varying the overall brightness of the picture; (5) a horizontal-hold control, which adjusts the horizontal deflection generator so that it conforms exactly to the control of the horizontal synchronizing impulses; and (6) a vertical-hold control, which performs the same function for the vertical deflection generator. Additional adjustments provided for the serviceman include adjustments for the height and width of the picture (magnitudes of the deflection currents) and adjustments (linearity controls) to secure precisely uniform rates of horizontal and vertical scanning to avoid distortion of the shapes of the objects in the picture. Colour receivers have additional controls to allow for adjustment of hue and intensity.

Television picture tubes. A typical television picture tube is illustrated in Figure 13. The tube is a highly evacuated, funnel-shaped structure. The viewing screen is located inside the tube face, a slightly curved glass plate that closes the wide end of the funnel. The screen itself, in monochrome receivers, is typically composed of two fluorescent materials, such as silver-activated zinc sulfide and silver-activated zinc cadmium sulfide. These materials, known as phosphors, glow with blue and yellow light, respectively, under the impact of high-speed electrons. The phosphors are mixed, in a fine dispersion, in such proportion that the combination of yellow and blue light produces white light of slightly bluish cast. A water suspension of these materials is settled on the inside of the face plate of the tube during manufacture, and this coating is overlaid with a film of aluminum sufficiently thin to permit the bombarding electrons to pass without

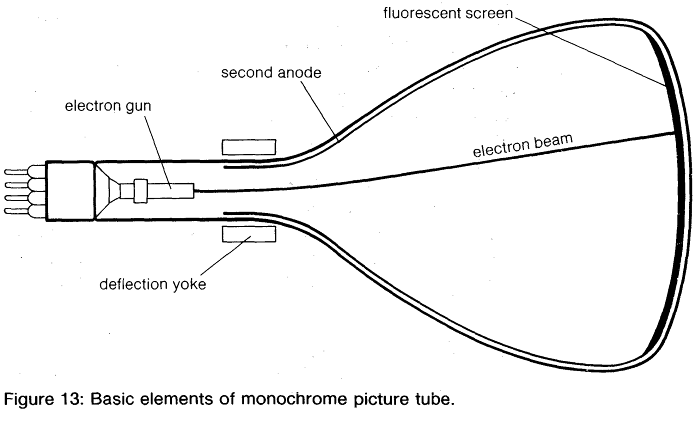

Figure 13: Basic elements of monochrome picture tube.

hindrance. The aluminum provides a mirror surface that prevents the backward-emitted light from being lost in the interior of the tube and reflects it forward to the viewer.

At the opposite end of the tube is the electron gun, a cylindrical metal structure that generates and directs a stream of free electrons, the electron beam.

The electron gun

A part of this gun is a cylindrical sleeve made electrically positive with respect to the cathode (the element that emits the electrons). The positively charged sleeve (first anode) draws the negative electrons away from the cathode, and they move down the sleeve toward the viewing screen at the opposite end of the tube. They are intercepted, however, by a flat disk (the control electrode) having a small circular aperture at its centre. Some of the moving electrons pass through the aperture; others are held back.

The television picture signal is applied between the control electrode and the cathode. During those portions of the picture signal wave form that make the potential of the control electrode less negative, more electrons are permitted to pass through the control aperture, whereas during the more negative portions of the wave, fewer electrons pass. The brightness control applies a steady (but adjustable) voltage between control electrode and cathode. This voltage determines the average number of electrons passing through the aperture, whereas the picture signal causes the number of electrons passing through the aperture to vary from the average and thus controls the brightness of the spot produced on the fluorescent screen.

As the electrons emerge from the sleeve, each electron experiences a force that directs it toward the centre of the viewing screen. The electron beam is thus brought to focus on the screen, and the light there produced is the scanning spot. Additional focussing may be provided by an adjustable permanent magnet surrounding the neck of the tube. The scanning spot must be intrinsically very brilliant, since (by virtue of the integrating property of the eye) the light in the spot is effectively spread out over the whole area of the screen during scanning.

Scanning is accomplished by two sets of electromagnet coils. These coils must be precisely designed to preserve the focus of the scanning spot no matter where it falls on the screen, and the magnetic fields they produce must be so distributed that deflections occur at uniform velocities.

The current in the horizontal deflection coil must change rapidly in order to provide a rapid retrace of the scanning spot on the face of the picture tube. This rapid rate of change of current causes pulses of high voltage to appear across the circuit feeding current to the coil, and the succession of these pulses, smoothed into direct current by a rectifier tube, serves as the high-voltage power supply.

The two sets of deflection coils are combined in a structure known as the deflection yoke, which surrounds the neck of the picture tube at the junction of the neck with the funnel section.

Variation in picture tubes

Picture tubes vary widely in screen size and are usually characterized by measurements made diagonally across the tube face. Tubes having diagonals from as small as three inches (7.5 centimetres) to 18 inches (46 centimetres) are used in typical transistor portable receivers, whereas tubes measuring from 23 to 27 inches (58 to 69 centimetres) are used in table- and console-model receivers. The design trend has called for greater angles of deflection, with correspondingly wider funnel sections and shallower overall depth from electron gun to viewing screen. The increase in deflection angle from 55° in the first (1946) models to 114° in models produced nowadays has required corresponding refinement of the deflection system because of the higher deflection currents required and the greater tendency of the scanning spot to go out of focus at the edges of the screen.

COMPATIBLE COLOUR TELEVISION

Basic principles—United States system. The technique of compatible colour television utilizes two transmissions. One of these, the luminance (brightness) transmission, employs methods essentially identical to those of the monochrome television system. The second, the chrominance (colour) transmission, has virtually no effect on monochrome receivers. When used with the luminance transmission in a colour receiver, however, it produces an image in full colour.

The luminance–chrominance method of representing colour values is one of several alternative ways in which coloured light may be analyzed and synthesized. This method is particularly appropriate in a television system since it produces a compatible signal that can serve both black-and-white and colour receivers by the same broadcast.

Historically, compatibility was of great importance because it allowed colour transmissions to be introduced without obsolescence of the many millions of monochrome receivers in use. In a larger sense, the luminance–chrominance method of colour transmission is advantageous because it utilizes the limited channels of the radio spectrum more efficiently than other colour transmission methods.

To create the luminance–chrominance values, it is necessary (in the present state of technology) first to analyze each colour in the scene into primary colours. Coloured light may thus be analyzed by passing the light through three coloured filters, typically red, green, and blue. The amounts of light passing through each filter, plus a description of the colour transmission properties of the filters, serve uniquely to characterize the coloured light.

The fact that virtually the whole range of colours may be synthesized from only three primary colours is essentially a description of the process by which the eye and mind of the observer recognize and distinguish colours. Like visual persistence, this is a fortunate property of vision, since it permits a simple three-part specification to represent any of the 10,000 or more colours and brightnesses that may be distinguished by human vision. If vision were dependent on the energy versus wavelength relationship (the physical method of specification), it is doubtful if colour reproduction could be incorporated in any mass-communication system.

By transforming the primary-colour values, it is possible to specify any coloured light by three other numbers: (1) its luminance (brightness or "brilliance"); (2) its hue (the redness, orangeness, blueness, or greenness, etc., of the light); and (3) its saturation (its vivid versus pastel quality). If the intended luminance value of each point in the scanning pattern is transmitted by the methods of the monochrome television system, it is only necessary to transmit, via an additional two-valued signal, supplementary information giving the hue and saturation of the intended colour at the respective points in the scanning pattern.

Chrominance, defined as that part of the colour specification remaining when the luminance is removed, represents two independent quantities, hue and saturation.

The chrominance signal

In the colour television system of the United States, the chrominance signal is an alternating current of precisely specified frequency (3.579545 ± 0.000010 megahertz), the precision permitting its accurate recovery at the receiver even in the presence of severe noise or interference. Any change in the amplitude of its alternations at any instant corresponds to a change in the saturation of the colours being passed over by the scanning spot at that instant, whereas a shift in time of its alternations (a change in the "phase angle" of the alternations) similarly corresponds to a shift in the hue. The colour information in the European (PAL and SECAM) systems is carried on a chrominance signal frequency of 4.43+ megahertz. In the United States system, as the different saturations and hues along each scanning line are successively uncovered by scanning in the camera, the amplitude and phase, respectively, of the chrominance signal change accordingly. The chrominance signal is thereby simultaneously modulated in amplitude and in phase. This doubly modulated signal is imposed on the picture signal carrier current, along with the luminance signal.

The television channel, when occupied by such a compatible colour transmission, appears as shown in Figure 14 (compare this with the channel when occupied by a black-and-white transmission, Figure 11). The chrominance signal takes the form of a subcarrier located precisely 3.579545 megahertz from the picture carrier frequency.

The picture carrier is thus simultaneously amplitude-modulated by the luminance signal to represent changes

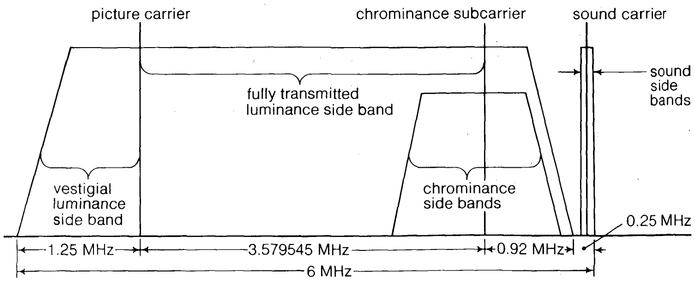

Figure 14: Allocation channel for compatible colour transmission in the United States.

in the intended luminance, and by the chrominance subcarrier that, in turn, is amplitude-modulated to represent changes in the intended saturation and phase-modulated to represent changes in the intended hue.

When compatible colour transmissions are received on a monochrome receiver, the receiver treats the chrominance subcarrier as though it were a part of the intended monochrome transmission. If steps were not taken to prevent it, the subcarrier would produce interference in the form of a fine dot pattern. The dot pattern, fortunately, can be rendered almost invisible in monochrome reception by deriving the timing of the scanning motions directly from the source that establishes the chrominance subcarrier itself. The dot pattern of interference from the chrominance signal, therefore, can be made to have opposite effects on successive scannings of the pattern; that is, a point brightened by the dot interference on one line scan is darkened an equal amount on the next scan of that line, and the net effect of the interference, integrated in the eye over successive scans, is virtually zero. Thus, the monochrome receiver, in effect, ignores the chrominance component of the transmission. It deals with the luminance signal in the conventional manner, producing from it a monochrome image. This monochrome rendition, incidentally, is not a compromise; it is essentially identical to the image that would be produced by a monochrome television system viewing the same scene.

The channel for colour transmissions, when used by colour receivers, would appear to be affected by mutual interference between the luminance and chrominance components, since these occupy a portion of the channel in common. Such interference is avoided by the fact that the chrominance subcarrier component is rigidly timed to the scanning motions. The luminance signal, as it occupies the channel, is actually concentrated in a multitude of small spectrum segments, by virtue of the periodicities associated with the scanning process. Between these segments are empty channel spaces of approximately equal size. The chrominance signal, arising from the same scanning process, is similarly concentrated. Hence it is possible to place the chrominance channel segments within the empty spaces between the luminance segments, provided that the two sets of segments have a precisely fixed frequency relationship. The necessary relationship is provided by the direct control by the subcarrier of the timing of the scanning motions. This intersegmentation is known as frequency interlacing. It is one of the fundamentals of the compatible colour system, and without it the superposition of colour information on a channel originally devised for monochrome transmissions would not be feasible.

Frequency interlacing

When a colour receiver is tuned to the transmission represented in Figure 14, the picture signal is recovered in a video detector in the usual manner. An amplifier stage, tuned to the 3.58-megahertz chrominance frequency, then selects the chrominance from the picture signal and passes it to a detector that recovers independently the amplitude-modulated and phase-modulated components.

European colour systems. The United States colour system has been adopted by Canada, Mexico, Japan, and several other countries. In Europe, two alternative systems have been developed and introduced. The PAL (phase alternation line) system has been adopted in the United Kingdom and most of the countries on the Continent. France, the Soviet Union, and their allied countries and dependencies have adopted the SECAM (*système électronique couleur avec mémoire*) system.

The PAL and SECAM systems embody the same principles as the United States system, including matters affecting compatibility and the use of a separate signal to carry the colour information at low detail superimposed on the high-detail luminance signal. The European systems were developed, in fact, to improve on the performance of the United States system in only one area, the constancy of the hue of the reproduced images.

It has been pointed out that the hue information in the U.S. system is carried by changes in the phase angle of the chrominance signal, and that these phase changes are recovered in the receiver by synchronous detection. The transmission of the phase information, particularly in the early stages of colour broadcasting in the United States, was subject to incidental errors arising in broadcasting stations and network connections. Errors are also caused by reflections of the broadcast signals by buildings and other structures in the vicinity of the receiving antenna. In recent years, transmission and reception of hue information has become substantially more accurate in the United States through care in broadcasting and networking, as well as by automatic hue-control circuits in receivers. The PAL and SECAM systems are inherently less affected by phase errors.

In the PAL and SECAM systems, the nominal value of the chrominance signal is 4.43+ megahertz, a frequency that is derived from and hence accurately synchronized with the frame-scanning and line-scanning rates. This chrominance signal is accommodated within the six-megahertz range of the fully transmitted side band. By virtue of its synchronism with the line and frame scanning rates, its frequency components are interleaved with those of the luminance signal, so that the chrominance information does not affect reception of colour broadcasts by black-and-white receivers.

PAL system. The PAL system resembles the United States system in that the chrominance signal is simultaneously modulated in amplitude to carry the saturation (pastel versus vivid) aspect of the colours and in phase to carry the hue aspect. In the PAL system, the phase information is reversed during the scanning of successive lines. Thus if a phase error is present during the scanning of one line, a compensating error (of equal amount but in the opposite direction) is introduced during the next line. Thus the average phase information, presented by the two successive lines taken together, is free of error.

Two lines are thus required to depict the corrected hue information, and the vertical detail of the hue information is correspondingly lessened. This produces no serious degradation of the picture when the phase errors are not too great, since, as noted above, the eye does not require fine detail in the hues of colour reproduction, and the mind of the observer averages out the two compensating errors. If the phase errors are more than about 20°, visible degradation does occur. This effect can be corrected by introducing in the receiver (as in the SECAM system) a delay line and electronic switch.

SECAM system. In the SECAM system, the luminance information is transmitted in the usual manner, and the chrominance signal is interleaved with it. But the chrominance signal is modulated in only one way. The two types of information required to encompass the colour values do not occur concurrently, and the errors associated with simultaneous amplitude and phase modulation do not occur. Rather, in the SECAM system (SECAM III), alternate line scans carry information on luminance and red, while the intervening line scans contain luminance and blue. The green information is derived within the receiver by subtracting the red and blue information from the luminance signal. Since individual line scans carry only half the colour information, two successive line scans are required to obtain the complete colour information, and this halves the colour detail, measured in the vertical dimension. But, as noted above, the eye is not sensitive to the hue and saturation of small details, so no adverse effect is introduced.

To subtract the red and blue information from the luminance information to obtain the green information, the red and blue signals must be available in the receiver simultaneously, whereas they are transmitted in time sequence. The requirement for simultaneity is met by holding the signal content of each line scan in storage (or "memorizing" it, hence the name *système électronique couleur avec mémoire*). The storage device is known as a delay line; it holds the information of each line scan for 0.000064 second. To match successive pairs of lines, an electronic switch is also needed. When the use of delay lines was first proposed, such lines were expensive devices. Recent advances have reduced the cost, and the fact that receivers must incorporate these components is no longer viewed as decisive.

Since the SECAM system reproduces the colour information with a minimum of error, it has been argued that SECAM receivers do not require receiver controls for hue and saturation. Such adjustments, however, are usually provided to permit the viewer to adjust the picture to his individual taste and to correct for signals that are in error as broadcast, due to such factors as faulty use of cameras, lighting, and networking.

Colour camera and picture tube. The terminals of the colour television system, the camera and picture tube, analyze and synthesize, respectively, the colour quantities present at each point in the scanning pattern. The television camera in its current state of development does not produce the luminance, hue, and saturation values directly. Rather it produces three picture signals, representative of the amounts of three primary colours present at each point in the image pattern. From these signals the luminance and chrominance components are then derived by manipulation in electronic circuits.

Optical system of a colour television camera

The typical colour television camera (Figure 15) contains three image orthicon tubes, with an optical system that casts an identical image on the sensitive surface of each of the three tubes. The optics consist of a lens and four mirrors that reflect the image rays from the lens onto the three camera tubes, as shown. Two of the mirrors are of a colour-selective type (dichroic mirror) that reflects the light of one colour and transmits the remaining colours. The mirrors, augmented by colour filters that perfect their colour-selective action, direct a blue image to the first tube, a green image to the second tube, and a red image to the third tube.

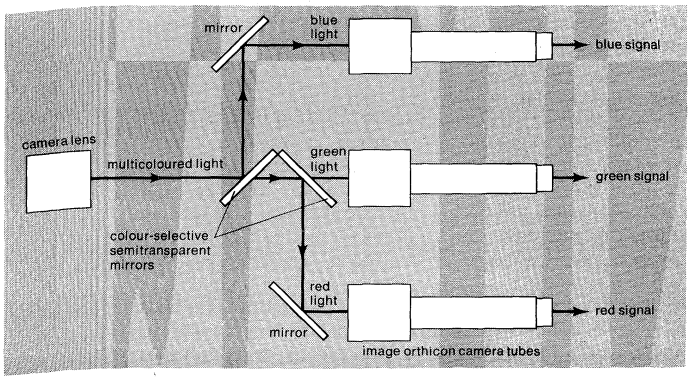

Figure 15: Diagram showing selective separation of light in colour camera.

The deflection systems of the three camera tubes are arranged to produce identical scanning patterns in each tube, so the picture signals developed by the respective tubes represent images of the same geometrical shape, differing only in colour. The respective primary-colour signals are passed through video preamplifiers associated with each camera tube and emerge from the camera as separate entities. If compatibility were not required, and if sufficient space were available in the radio spectrum, the three signals might be transmitted to the receiver over separate channels. Actually, as mentioned above, the primary-colour signals are recast into luminance and chrominance components; at the receiver the latter are converted back into primary-colour signals before application to the colour picture tube. In the mid-1960s, colour cameras employing four camera tubes were introduced, particularly for studio work. The fourth tube is exposed directly to the scene without an intervening mirror or filter and thus produces a luminance signal of high quality. The three remaining tubes are then confined to the less demanding functions of producing the primary-colour signals. Three-tube cameras based on this principle have also been developed.

Much attention has been paid to the design of camera tubes for colour service, and greatly improved colour rendition, particularly of subtle variations in skin tone, has resulted. The Plumbicon version of the Vidicon, developed in The Netherlands, is particularly noteworthy in this respect.

The colour picture tube (Figure 16) contains three electron guns, which produce three separate electron beams. These are deflected simultaneously in the standard scanning pattern over the viewing screen. One of the beams is controlled by the red primary-colour signal and produces a red image. The second beam produces a blue image, and the third, a green image.

Colour picture tube screen

To permit three primary-colour images to be formed simultaneously, the screen is composed of three sets of individual phosphor dots, which glow respectively in the three different colours and which are uniformly interspersed over the screen (Figure 16). The "red" beam impinges only on the red-glowing phosphor dots, being prevented from hitting the other two colours, with similar restriction of the other two beams to the blue and green dots, respectively.

The sorting out of the three beams so that they produce images of only the intended primary colour is performed by a mask that lies directly behind the phosphor screen. This mask contains about 200,000 precisely located holes, each accurately aligned with three different coloured phosphor dots on the screen in front of it. Electrons from the three guns pass together through each hole, but each electron beam is directed at a slightly different angle. The angles are such that the electrons arising from the gun controlled by the red primary-colour signal fall only on the red dots, being prevented from hitting the blue and green dots by the shadowing action of the mask. Similarly, the "blue" and "green" electrons fall only on the blue and green dots, respectively. A major improvement consists in surrounding each colour dot with an opaque black material, so that no light can emerge from the portions of the screen between dots. This permits the screen to produce a brighter image while maintaining the purity of the colours. This type of screen (black matrix screen) is more expensive to produce, but it has been widely used in larger colour tubes.

Thus three separate primary-colour images are formed simultaneously, as the scanning proceeds. The colour dots of which each image is formed are so small and so uniformly dispersed that the eye does not detect their separate presence, although they are readily visible through a

By courtesy of RCA

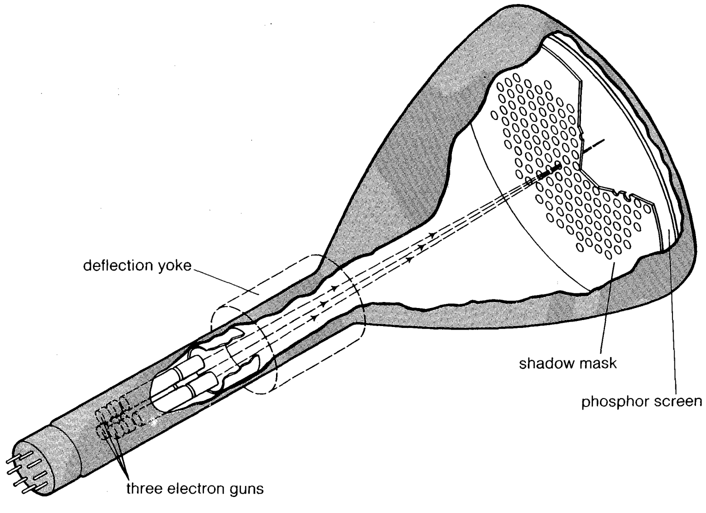

Figure 16: Colour picture tube employing the shadow-mask technique.

magnifying glass. The primary colours in the three images thereby mix in the mind of the viewer and a full-colour rendition of the image results.

The shadow-mask tube

This type of colour tube is known as the shadow-mask tube. It has several shortcomings: (1) the electrons intercepted by the mask cannot produce light, and the image brightness is thereby limited; (2) great precision is needed to achieve correct alignment of the electron beams, the mask holes, and the phosphor dots at all points in the scanning pattern; and (3) precisely congruent scanning patterns, as among the three beams, must be produced. In the late 1960s a different type of colour tube, the Trinitron, was introduced in Japan. In this tube the shadow-mask is replaced by a metal grille having vertical slits extending from the top to the bottom of the screen. The three electron beams pass through the slits to the coloured phosphors, which are in the form of vertical stripes aligned with the slits. The grille directs the majority of the electrons through the slits, causing a much lower percentage of the electrons to be intercepted by the grille, and a brighter picture results. The electron gun of the Trinitron tube remains in fine focus throughout the scanning of the image, since a major portion of the focus mechanism is common to all three beams.

Colour transmitter and receiver. Figure 17 shows the essential elements of the colour transmitter, as they differ from the monochrome transmitter shown in Figure 12.

By courtesy of RCA

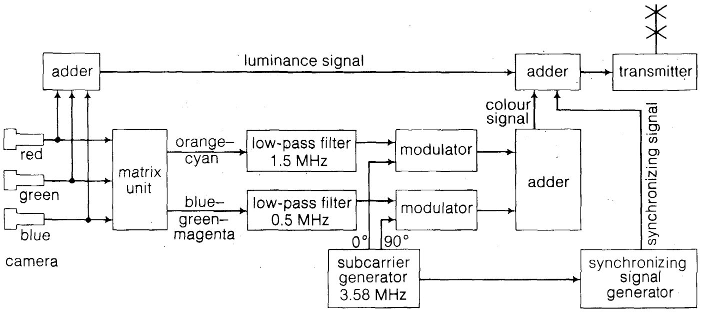

Figure 17: Block diagram of colour transmitter.

Immediately following the colour camera is the colour coder, which converts the primary-colour signals into the luminance and chrominance signals. The luminance signal is formed simply by applying the primary-colour signals to an electronic addition circuit that adds the values of the three signals at each point along the respective picture signal wave form. Since white light results from the addition, in appropriate proportions, of the primary colours, the resulting sum signal represents the black-and-white (luminance) version of the colour image. The luminance signal thus formed is subtracted individually, in three electronic subtraction circuits, from the original primary-colour signals, and the colour-difference signals are then further combined to produce two signals, which are then applied simultaneously to a modulator, where they are mixed with the chrominance subcarrier signal. The latter is thereby amplitude-modulated in accordance with the saturation values and phase-modulated in accordance with the hues. The luminance and chrominance components are then combined to form the overall colour picture signal, which is then carried to the transmitter over the network or studio–transmitter link.

The chrominance subcarrier is generated in a precise electronic oscillator at the standard value of 3.579545 megahertz. Samples of this subcarrier are injected into the signal wave form during the blank period between line scans just after the horizontal synchronizing pulses. These samples are employed in the receiver to control the synchronous detector, mentioned above. Finally, horizontal and vertical deflection currents, which produce the scanning in the three camera tubes, are formed in a scanning generator, the timing of which is controlled by the chrominance subcarrier. This common timing of deflection and chrominance transmission produces the dot-interference cancellation in monochrome reception and frequency interlacing in colour transmission, noted above.

At this stage, the colour picture signal can be handled in the same manner as the monochrome signal, except that throughout the network and in the transmitter special precautions must be taken to transmit the chrominance subcarrier at its correct amplitude, with its amplitude and phase modulations preserved against incidental distortions that would result in poor colour quality.

Colour receiver components

The colour receiver contains a tuner and intermediate-frequency amplifiers similar to those used in monochrome receivers, but its components are especially designed to preserve the chrominance subcarrier during its passage through these stages. The colour receiver also has power supplies, sound reception circuits and synchronization, and deflection current generators similar to those used in monochrome reception.

The colour reproduction processes are divided into the luminance and chrominance functions. A video detector develops the luminance component and applies it through video amplifiers simultaneously to all three electron guns of the picture tube. This part of the signal thereby activates all three primary-colour images, simultaneously and identically, in the fixed proportion needed to produce white light. When tuned to monochrome broadcasts, the colour receiver produces a monochrome image by means of this mechanism, the chrominance component then being absent.

When the receiver is tuned to a colour broadcast, the chrominance subcarrier component appears in the output of the video detector and it is thereupon operated on in circuits that ultimately recover the primary-colour signals originally produced by the colour camera.

The three colour-difference signals are applied separately to the respective electron guns, where each reduces the strength of the corresponding electron beam to change the white light, which would otherwise be produced, to the intended colour for each point in the scanning line. The net control signal applied to each electron gun bears a direct correspondence to the primary colour signal derived from the respective camera tube at the studio. In this manner, the three primary-colour signals are transmitted as though three separate channels had been used.

The colour receiver controls to be adjusted by the viewer (in addition to the tuning, brightness, contrast, and synchronizing controls present in monochrome receivers) are the hue control and the saturation control. The hue control shifts all the hues in the reproduced image. It is usually adjusted by reference to the colour of the flesh of the performers, since unnatural flesh tone is readily recognized even if the viewer has no knowledge of the intended hues in other parts of the scene. The saturation control adjusts the magnitudes of the colour-difference signals applied to the electron guns of the picture tube. If these magnitudes are reduced to zero, by turning the saturation control to the "off" position, no colour difference action occurs and the reproduction occurs in monochrome. As the saturation control is advanced, the colour differences become more accentuated, and the colours become progressively more vivid. In the late 1960s, the more elaborate colour television receivers began to employ a system known as "automatic hue control." In this system, the viewer makes an initial manual adjustment of the hue control to produce a rendition of flesh tones to suit his preference. Thereafter, the hue control circuit automatically maintains the preselected ratio of the primary colours corresponding to the viewer's choice. Thus the most critical aspect of the colour rendition, the appearance of the faces of the performers, is prevented from changing when cameras are switched from scene to scene or when tuning from one broadcast to another.

Automatic hue control

Since care in tuning is essential in colour receivers to assure correct reception of the chrominance signal, automatic circuits to set the fine-tuning adjustment are provided in the more expensive receivers. Another development is a single touch-button control that sets the fine tuning and adjusts the hue, tint, contrast, and brightness to preset ranges. These automatic adjustments override the settings of the corresponding knobs, which then function over narrow ranges only. Such refinements permit reception of acceptable quality by viewers who might otherwise be confused by the many misadjustments possible when ordinary manual controls are used.

TELEVISION RECORDING

Recording of television programs on photographic film or magnetic tape is an important technique, not only to preserve a permanent record of a live-scene program for subsequent rebroadcast but also to perfect a performance prior to its initial presentation. Recorded segments are often used in a "live" studio program when it is impractical to set up the desired scene in the studio. Transmission of films originally produced for motion-picture theatres is also a major activity in broadcasting.

The continuous projector. Television film projectors fall into two classes, continuous and intermittent, according to the type of film motion. In the continuous projector, a scanning spot from a flying spot camera tube is passed through a rotating optical system, known as an immobilizer, which focusses the scanning spot on the motion-picture film. As the film moves continuously through the projector, the immobilizer causes the scanning pattern as a whole to follow the motion of the film frame, so that there is no relative motion between pattern and frame. The light passes through the film to a phototube where the light, modified by the transmissibility of the film at each point, produces the picture signal. As one film frame moves out of the range of the immobilizer, the next moves into range and there is a condition of overlap between successive scanning patterns.

The optics are so arranged that the amount of light in the spot focussed on the film is constant at all times and in all positions. This constancy permits the film to be moved at any desired speed, while the pattern scans at the standard rate of 25 or 30 pictures per second. The film is actually moved at the standard rate for motion pictures, 24 frames per second, so the speed of objects and pitch of the accompanying sounds (picked up from the sound track by conventional methods) are reproduced at the intended values.

The intermittent projector. In the intermittent projector, which more nearly resembles the type used in theatre projection, each frame of film is momentarily held stationary in the projector while a brief flash of light is passed through it. The light (which passes simultaneously through all parts of the film frame) is focussed on the sensitive surface of a storage-type camera, such as the Vidicon. The light flashes are timed to occur during the intervals between successive field scans—that is, while the extinguished scanning spot is moving from the bottom to the top of the frame. The light is strong enough to produce an intense electrical image in the tube during this brief period. The electrical image is stored and then is scanned during the next scanning field, producing the picture signal for that field. Light is again admitted between fields, and the stored image is scanned thereupon by the second field. When one film frame has been thus scanned, it is pulled down by a claw mechanism and the next frame takes its place.

In Europe and other areas where the television scanning rate is 25 picture scans per second, it has been the custom to operate intermittent projectors also at 25 frames per second, or about 4 percent faster than the intended film projection rate of 24 frames per second. The corresponding increases in speed of motion and sound pitch were not so great as to introduce unacceptable degradations of the performance. In the United States and other areas where television scanning occurs at 30 frames per second, it is not feasible to run the film projector at 30 film frames per second, since this would introduce speed and pitch errors of 25 percent. Fortunately, a small common factor, 6, relates the scan rate of 30 and the film projection rate of 24 frames per second. That is, four film frames consume the same time as five scanning frames. Thus, if four film frames pass through the projector while five complete picture scans (10 fields) are completed, both the film motion and the scanning proceed at the standard rates The two functions are kept in step by holding one film frame for three scanning fields, the next frame for two scans, the next for three scans, and so on.

Magnetic video recording. Television program recording for rebroadcast is done on magnetic tape, although film is used for longer storage purposes. The tape technique permits prompt rebroadcast of programs the quality of which, when the tape equipment is in good adjustment, is virtually indistinguishable from the original broadcast. The magnetic tape, two inches (five centimetres) wide, is moved at 15 inches (38 centimetres) per second past a magnetic recording head that rotates across the width of the tape, imposing on it a series of crosswise magnetic line patterns that correspond to the lines of the scanning pattern. To keep the pickup and tape motions identical during recording and playback, index signals are imposed during recording, and these are used to control the drive motors during playback. The television signal is recorded by frequency modulation on a carrier signal, thus avoiding the difficult problem of direct recording at the extremely low frequencies present in the picture signal itself.

Magnetic video recording is widely used in broadcast operations, in colour as well as monochrome, not only because of its high quality but because it is immediately available (without subsequent processing) for playback. In the United States, many of the network programs broadcast in and west of Chicago are derived from magnetic recordings, stored for an hour or more after their release along the eastern seaboard to allow for the difference in time in western cities.

Special techniques. Many variations of the basic techniques of recording television program material were developed in sports telecasting. The first to be introduced was the "instant replay" method, in which a video tape recording is made simultaneously with the live-action pickup. When a noteworthy episode occurs, the live coverage is interrupted and the tape recording is broadcast, followed by a switch back to live action. Often the recording is made from a camera viewing the action from a different angle. Other variations include the slow-motion and stop-action techniques, in which magnetic recording plays the basic role.

Increasing use has been made, particularly in sports broadcasting, of split-screen techniques and the related methods of inserting a portion of the image from another camera into a circular (or rectangular, triangular, etc.) area cut out from the main image. These techniques employ an electronic switching circuit that turns off the signal circuit of one camera for a portion of several line scans, while simultaneously turning on the signal circuit of another camera, the outputs of the two cameras being combined before the signal is broadcast. The timing of the electronic switch is adjusted to blank out, on successive line scans of the first camera, an area of the desired size and shape. The timing may be shifted during the performance and the area changed accordingly. One example of this technique is the wipe, which removes the image from one camera while inserting the image from another, with a sharp, moving boundary between them. (D.G.F./Ed.)

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Brussels

Brussels (Dutch: Brussel; French: Bruxelles), the capital of the Kingdom of Belgium and of the Belgian province of Brabant, consists of an agglomeration of about 1,000,000 inhabitants. It is divided into 19 communes, or municipalities, each with a large measure of administrative autonomy. Although the average visitor might remain unaware of this dispersal of local government powers, it holds the key to an understanding of many of the urban problems afflicting the contemporary city. It has, for example, impeded the harmonious development of the spreading built-up areas that reflect the increase in population that has occurred since 1830, and an adequate solution to the traffic problem has been rendered especially difficult.

A further major problem arises from the geographic location of the city. It lies just a few miles north of the invisible "language boundary" separating the nation's Dutch-speaking region of Flanders in the north from the southern, French-speaking Walloon region. Brussels is thus surrounded by Flemish territory and was historically a predominantly Dutch-speaking city, but at present the majority of residents in the Brussels agglomeration speak French. Officially, as the Belgian capital, the city is strictly bilingual, and in all spheres of public life Dutch and French are used side by side. Nevertheless, increasingly in the 20th century Brussels has been the principal venue for political clashes between Flemings and Walloons. Partly as a result of these conflicts, the Belgian Parliament reorganized the nation's structure on the basis of three regions: Flanders, Wallonia, and Brussels. Although the city acquired a separate political identity, its exact place and role in Belgium's decentralization measures have not been defined by law.

Brussels is nevertheless the administrative, commercial, and financial heart of Belgium, and all services and institutions of national importance are based in the city. Brussels is, in addition, a major European tourist and cultural attraction, functioning simultaneously as regional metropolis, national capital, and international centre. The last-named role has flourished since the city became host to the European Communities (made up of the European Economic Community [EEC], or Common Market; the European Coal and Steel Community; and the European Atomic Energy Community) as well as to the North Atlantic Treaty Organization (NATO) headquarters.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

Brussels lies in the central plateaus of Belgium, in the valley of the Senne River (Dutch: Zenne), a small tributary of the Scheldt (French: Escaut). Its relative proximity to the North Sea gives it a mild, moderate climate, with summer temperatures usually between 68° and 77° F (20° and 25° C) and winter temperatures rarely falling below 32° F (0° C).

Inner Brussels

Seen from the air, the historic Old Town of Inner Brussels has the shape of a pentagon. It forms the centre of the modern metropolis, but the walls that once surrounded it were replaced by a ring of boulevards in the mid-1800s, around the time that Belgium became an independent kingdom. Since then, Brussels has been transformed, in the Old Town as well as in the surrounding communes. The determining factor in this metamorphosis has been incessant population pressure, which caused a building boom and the development of an ever-widening network of streets, avenues, and roads crisscrossing the countryside and urbanizing the neighbouring villages.

At first the urban tentacles only pushed forward along the seven or eight routes radiating from the tollgates along the old city walls, but after the toll system was abolished in 1860 they also spread along new roads. The suburbs expanded rapidly beyond the town gates, and by the end of the 19th century the territory of several of the first ring of communes was completely or almost completely covered by residential buildings. The expansion continued into the 20th century, in all directions: north and south along the valley of the Senne, and east and west on the undulating plateaus separated by the tributaries of the Senne (Maalbeek, Woluwe, and others).

As a result, the landscape was entirely changed. With rare exceptions, the ponds and pools of earlier days were filled in; most of the hollows were banked up; the rivers and streams covered over and converted into sewers; the small woods cut down; fields, pastures, and orchards parceled out for development; and footpaths eliminated. The Senne River became a main sewer running under the wide, straight central boulevards of Inner Brussels, linking the North (1841) and the South (1869) railway stations. In 1911 the city began building a series of railway tunnels and viaducts connecting the North and South stations by way of the underground Central Station (1952). This so-called North–South link was completed in 1956. Although it has facilitated transportation, several decaying residential areas had to be demolished during its construction, an action that has since caused some regret.

Greater Brussels is not entirely built over, mainly because, with the introduction of rapid means of transport, urban development has leapfrogged to satellite communities rather than proceeding by continuous, uninterrupted expansion. Since the 19th century, the boundaries of the Brussels agglomeration changed repeatedly, giving the sprawling metropolis the appearance of an amoeba. In the latter third of the 20th century, however, legislation resulting from Belgium's linguistic and community conflicts strictly confined the city within the limits of its 19 constituent municipalities.

The Grand' Place

Inner Brussels is divided between the commercial quarter and the upper town, where the principal governmental buildings are situated. The commercial quarter extends from the western outer boulevards to a little east of the central boulevards and includes the medieval marketplace known as the Grand' Place. This square, with its elaborately decorated 17th-century guildhalls, lies at the heart of the Old Town. It is occupied on its south side by the imposing Town Hall and on its north by the ornate King's House (Maison du Roi; almost entirely rebuilt during 1873–95), which contains the historical museum.

The Town Hall in the Grand' Place, Brussels, built 1402–54.
By courtesy of Sabena Belgian World Airlines

One of the curiosities of this quarter is the Manneken-Pis Fountain (1619), noted for a small bronze statue of a boy urinating and known to the people of Brussels as their oldest "citizen."

The upper town is the remaining eastern area of the inner city. It is crossed from southwest to northeast by a major thoroughfare, on which stand the Royal Palace and the Palace of the Nation. The latter was erected (1779–83) by the Austrian governors and after independence became the home of the Senate and the Chamber of Representatives. It stands at the intersection of the rue Royale (Koningsstraat) and the rue de la Loi (Wetstraat), an area that has become a symbol for the national government.

THE PEOPLE

The population of the Brussels agglomeration grew steadily from 57,000 in 1755, when the first census was held, to 104,000 in 1830, 626,000 in 1900, and 892,000 in 1930. It reached a peak of a little more than 1,000,000 in 1970 but has since declined to just under 1,000,000. The population of the inner city increased in line with that of the total agglomeration until around 1890, when it stood at 160,000, but it has declined sharply since the beginning of the 20th century, falling to 141,000 in 1910, 84,000 in 1947, and around 60,000 in the 1960s.

The effect of immigration

Immigration has had a significant impact on the demographic and linguistic evolution of the city. In the 19th century the immigrants usually came from Flanders or Wallonia, although there was also a large expatriate community from France and, to a lesser extent, Germany. Until then Brussels remained the Flemish city it had always been, with only about one-third of its inhabitants speaking French. The new Flemish immigrants, however, belonged on the whole to the lower strata of society (domestic servants, labourers), whereas the Walloon immigrants were predominantly middle-class employees and civil servants. Largely as a result of social pressure and the prestige of French, by the mid-20th century a large majority of Bruxellois spoke French. No language census has been taken since the controversial 1947 census, the matter being politically too sensitive, but it is thought that at least three-quarters of the population of Brussels may now be predominantly French-speaking.

In many respects Dutch and French speakers have their own cultural circuits. Broadly speaking, Dutch-language cultural life is more in evidence in the northwestern part of the agglomeration and French-language culture in the southeastern part. A number of cultural establishments, however, from the Théâtre de la Monnaie (Muntschouwburg) to the Ancient Belgium cabaret, are shared by the two communities, many middle-class Bruxellois being conversant in both languages. In recent years a new Flemish elite has emerged, improving the prospects for the Dutch language in the Belgian capital.

Since the mid-1960s, immigration of EEC personnel from western European countries and, in much larger numbers, of manual workers and their families from various Mediterranean countries has substantially altered the composition of the population of Brussels. Whereas in 1961 foreigners accounted for less than 7 percent of the population, by 1985 their share was more than 25 percent. EEC employees generally live in the upper part of the city or in the leafier suburbs. The immigrants from the Mediterranean—from Spain, Italy, and Greece at first, but later also from Morocco and Turkey—form a much larger group in central Brussels. A young population with a high birth rate, they make up as much as half the population of such boroughs as Saint-Josse-ten-Node (Sint-Joost-ten-Node) and Saint-Gilles (Sint-Gillis). In general, these immigrants retain their cultural distinctness; the municipality of Schaerbeek (Schaarbeek), for example, has 18 mosques for its foreign residents.

THE ECONOMY

In the early days of national independence, in the 1830s, food markets in the Old Town were supplied by the neighbouring villages. These have since been urbanized, and industrial, commercial, and service activities have been substituted for rural work.

Brussels' most important industrial zone is located on the city's south–north axis, along the valley of the Senne, where the port of Brussels developed and where the railway lines run. Food processing and the manufacture of machinery, electrical products, chemicals, and textiles are the leading industries. White-collar workers far outnumber blue-collar workers, however. In fact, the financial and service industries account for more than two-thirds of all employment in Brussels.

Economic and financial importance of the city

The capital has been the financial heart of Belgium and a major commercial centre ever since the private and powerful holding company the Société Générale de Belgique was established there in 1822. Domestic and foreign banking institutions and insurance companies have appeared in increasing numbers. Because most of the large Belgian industrial and commercial concerns have their registered offices in Brussels, it constitutes the decision-making centre for economic and financial affairs. With the establishment of the European Common Market, many multinational corporations have set up their regional coordinating offices there. Brussels has also developed into an international economic centre, with an important stock exchange and an annual commercial fair.

Brussels' working population amounts to almost half of its total population; this high proportion is explained by the fact that many women are employed, usually in the commercial or service sectors. Both the standard and the cost of living are relatively high; the average income per person is significantly higher in Brussels than in either Flanders or Wallonia, although the gap is narrowing.

The importance of Brussels to the national economy is evident from the fact that, with 10 percent of the country's population, the city provides work for 17 percent of Belgium's economically active population. Commuters make up one-third of the entire Brussels work force; more than two-thirds of the commuters come from Flanders, mostly from Flemish Brabant. As a result, the streets of the city change dramatically at rush hours, when tens of thousands of commuters are traveling, predominantly by train and car. To alleviate the worsening traffic problems and to counter the diminishing use of communal transport in the agglomeration, in 1965 the city began developing a comprehensive subway network. The first line was opened in 1976; five lines now extend to all parts of Greater Brussels.

The city has also developed an extensive external transportation system. A ring of modern motorways surrounds Greater Brussels, forming a hub of radiating highways that link the major cities of Belgium. Brussels is also the focal point of the Belgian railway system, one of the densest in the world. At peak times the city's Central Station, which is restricted to six underground tracks, has up to 100 trains passing through every hour. A special line connects the Central Station with the Brussels National Airport (at Zaventem, to the east) in approximately 10 minutes.

The city's easy accessibility and central geographic location in western Europe have proved beneficial to its tourist trade. Many overseas visitors to the Continent use Brussels as a convenient base or starting point for their travels.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Brussels was the historic capital of the duchy of Brabant; after Belgian independence it continued as the capital of the province of Brabant and as such houses the provincial assembly and the governor. The de facto national capital since the 15th century, Brussels became the de jure capital of the Kingdom of Belgium in 1830. The king has his palace there, facing the Palace of the Nation, which houses the Chamber of Representatives and the Senate. For a long time the various ministries were close by, but several have migrated to the Residence Palace and to other, more functional administrative buildings.

The Belgian constitution stipulates that matters concerning individual communes only are to be settled by the local council, according to specified principles. In 1836 Parliament passed the "organic" communal law, which provided for the autonomy of each commune. This explains why Greater Brussels for long was governed by 19 separate authorities and not by one single authority. Attempts to coordinate administration, such as the formation of the Assembly of Burgomasters in 1933, had no official status.

Constitutional change of 1970–71

In 1970–71 Parliament amended the constitution to provide for the recognition of three regions (Dutch-speaking Flanders, French-speaking Wallonia, and bilingual Brussels) as well as federations of communes. Each of the 19 communes of Brussels continued, however, to have its own council, municipal establishments, burgomaster, and alderman. The council in each commune is a deliberative assembly whose members are elected by universal suffrage; the aldermen are chosen by the councillors from among themselves. The burgomaster is appointed by the king, usually from among the councillors. He represents the head of state in the commune and sees to it that laws and regulations are carried out. He acts as local chief of police and in emergencies can take whatever steps are necessary to maintain or restore public order. The burgomaster of Brussels is a political figure of importance.

The 1970–71 constitutional revision also created a Council for the Brussels Agglomeration, a directly elected assembly with its own president and executive. The council was assigned specific powers in such matters as urban planning, economic expansion, environmental issues, fire protection, and ambulance service. New elections for the council, scheduled for 1976, were postponed indefinitely in view of impending further reforms of the Belgian state. These reforms were implemented in 1980, but because the legal definition of the powers to be awarded to the Brussels region was again put off indefinitely, the metropolis provisionally came under the control of a ministerial committee, which consists of government ministers and is answerable only to the national parliament. The 1980 legislation did, however, place cultural matters and matters relating to health care and welfare in Brussels under the jurisdiction of either the Flemish Council or the French Community Council, depending on whether they affect the city's Dutch-speaking or French-speaking inhabitants.

Public services. Brussels' communal services proliferated after 1830 as the city's population grew and became more mobile. The effect of larger scale and faster speed has meant growth in existing administrative departments as well as the creation of many new ones, such as the water, gas, and electricity administrations and the departments for youth and sports, for the aged, for burial services, and for education and the fine arts. Several of the communal industrial administrations have been replaced by intercommunal corporations (water, gas, and electricity). The public transport services (trains and buses) have been entrusted to a corporation of this kind.

Education. Approximately half of the elementary-school population attends so-called free (*i.e.,* Roman Catholic, but state-subsidized) schools. "State" schools are run by the local communes, the province of Brabant, or the Belgian state. Families are free to send their children to either Dutch-language or French-language schools. The reluctance of some municipalities to provide facilities for primary and secondary education in Dutch caused resentment among Flemings in the 1950s; even in the mid-1980s seven of the 19 communes of the agglomeration had no Dutch-language municipal schools. Inner Brussels has played an exemplary role in setting up scholastic institutions, most notably its generous contributions to the foundation in 1834 of the Free University of Brussels and to its development.

CULTURAL LIFE

Educational institutions

Brussels is an artistic and tourist centre, with a wide variety of cultural activities. In addition to the Free University—divided since 1970 into a French-speaking and a Dutch-speaking university—the royal academies of science, medicine, French language and literature, and Dutch language and literature, are based there, as are various other institutes of higher learning, including the largest branch of the National Archive, the Albert I Royal Library, and many museums of national or local importance.

The Palace of Fine Arts, designed by the Art Nouveau architect Victor Horta and opened in 1928, provides a cultural centre for those interested in the visual arts, film, music, literature, and the theatre. Most of the city's large-scale art exhibitions are presented there, and it is the headquarters of the Philharmonic Society and the National Federation of Youth and Music. The annual Queen Elisabeth of Belgium International Music Contest attracts worldwide interest. Midday poetry readings and concerts are held weekly. Most of the communes in the agglomeration have, on the model of the Palace of Fine Arts, established cultural centres that organize exhibitions, stage shows, and concerts.

French cultural influence

The two outstanding periods in Brussels' cultural history were the late medieval flowering under the Burgundians (most of the town's Gothic churches date from this era) and the late 19th to early 20th century, when Brussels was a centre of innovation in literature, theatre, architecture (Henry van de Velde, Horta), and painting (the Surrealists Paul Delvaux and René Magritte). Contemporary cultural life is cosmopolitan, although France provides much of the inspiration. Foremost among the theatres are the French-language National Theatre and the National Opera House. The French choreographer Maurice Béjart and his Ballet of the 20th Century were based in the city from 1960 until 1987. The minority Dutch-language culture of Brussels is less visible but quite dynamic.

Not far from the urban centre are scenic walks in the magnificent beech groves of the Soignes Forest (Zoniënwoud) and its offshoot, the Cambre (Terkameren) Woods. The city's main sports stadium is located in Heysel (Heizel), a northern district of the Brussels commune, where the 1958 World Exhibition was held.

History

EARLY SETTLEMENT AND GROWTH

Origins. The oldest known reference to Brussels dates to the 7th century and has the form Bruocsella, which means "settlement in the marshes." The name reflects the fact that the city owes its origin to the establishment, in the 6th century, of a fortified castle on a small island in the Senne River, which flows from south to north, and to an east–west economic route linking Rhenish towns such as Cologne with Brugge (Bruges), Ypres (Ieper), and other towns in the county of Flanders. At the point where road and river crossed, a market and bartering place developed under the protection of the dukes of Brabant. By the 12th century the settlement was surrounded by defensive ramparts with towers and fortified gateways.

During the 12th, 13th, and 14th centuries Brussels grew to become one of the major towns of the duchy of Brabant. Its economic mainstay was the manufacture of luxury fabrics, which were exported to fairs in Paris, Venice, the Champagne region of France, and elsewhere. The cloth trade made fortunes for a few enterprising merchant families, who developed into seven dynasties that, with the help of the duke of Brabant, acquired a position of complete political mastery. In control of business and municipal affairs, they also exercised power as magistrates, giving rulings on disputes arising among the inhabitants, as well as acting as a court of appeal for neighbouring areas. The prevailing regime was, in fact, strongly plutocratic in nature.

Abuse of such powers provoked violent popular uprisings in 1280, 1303, 1360, and 1421. This last upheaval led to a more equitable system of government, with local powers divided between the patrician families and the emergent guilds of craftsmen and other workers. Gradually, however, the patrician elite regained political control; as late as 1719 a popular revolt led by Frans Anneessens ended with his public execution.

Centuries of occupation. Events of particular significance in the 14th century were the invasion of the duchy of Brabant by the Count of Flanders' troops, their brief occupation of Brussels, and the construction, immediately after the town's liberation, of huge fortified walls (1357–79), which survived until the first half of the 19th century.

Burgundian era

In 1430 the duchy was merged in the possessions of the duke of Burgundy. The Burgundian period, which lasted until 1477, was one of political and artistic prestige. Brussels became the seat of the central administrative bodies for the ducal possessions in the Low Countries, which constituted a rich centre of art and culture. Pictures by Rogier van der Weyden, the officially appointed town painter, sculptures in wood, large tapestries with historical motifs, plate, jewelry, and other products by Brussels craftsmen came to be exported in all directions.

Brussels began to beautify itself: by the marketplace, the Town Hall (1402–54) rose proudly, with its tall perforated steeple surmounted by a statue of the archangel Michael, the city's patron saint. Various Gothic churches and cathedrals and the ducal Coudenberg Palace, with its extensive park, added to the architectural splendour.

After a prolonged political crisis caused by an abortive rebellion against the future Holy Roman emperor Maximilian I at the end of the 15th century, Brussels regained its position as a capital during the reign of Charles V (1519–59). The three government councils (the Council of State, the Privy Council, and the Finance Council) were established there permanently. The city's population grew to nearly 50,000 by the mid-16th century. In 1561 a canal linking Brussels with Willebroek was dug, providing direct access to the Rupel and the Scheldt rivers and thus to the port of Antwerp and the North Sea. Replacing the sandy little Senne river, the Willebroek Canal played an important commercial role.

The Reformation did not leave Brussels untouched. Two Lutheran preachers died there at the stake in 1523, the first Protestant martyrs in the Low Countries; many more Lutherans, Anabaptists, and Calvinists followed. During the Revolt of the Netherlands, Brussels was under Calvinist rule from 1578 until 1585, when the southern provinces of the Low Countries (which included modern-day Belgium) separated from the northern provinces (now the Kingdom of The Netherlands), surrendered to the Spanish Habsburgs, and returned to the Roman Catholic fold.

The Counter-Reformation and the reign of Archduke Albert and Isabella (1598–1633) left their mark on the urban surroundings with the construction of a series of fine churches in the Italo-Flemish Baroque style, nearly all of which are still in existence. In the second half of the 17th century there were repeated invasions by the armies of Louis XIV of France. During a bombardment by his troops in 1695, hundreds of buildings were destroyed by fire, including the various craft headquarters. Out of this catastrophe there arose new guildhalls, the architectural landmarks now surrounding the Grand' Place.

Brussels suffered a brief but costly occupation by French troops in 1746–48, but as part of the Austrian Netherlands it profited from the general economic recovery in the latter half of the 18th century, becoming a financial centre and gaining new industries. The upper part of the town was the scene of urban planning on a large scale, which resulted in the Place Royale and Brussels Park. The park is, following the French model, perfectly symmetrical and surrounded by Neoclassical buildings, of which the largest is the Palace of the Nation.

Following the Brabant revolt (1788–90) against the government of Emperor Joseph II of Austria, the French republican armies made their appearance, and the Belgian principalities were annexed to France. During the Napoleonic era, Brussels was reduced to the rank of chief town of the French *département* of the Dyle, losing in addition all authority over its satellite villages.

One of the consequences of Napoleon's defeat at Waterloo was the creation of the United Kingdom of The Netherlands. This reunion of the southern and northern provinces, which had been separated in the 16th century, lasted 15 years (1815–30). During this period Brussels shared the status of capital with The Hague. Its appearance changed appreciably, above all because of the demolition of the city walls (1810–40) and their replacement by tree-lined boulevards, as well as the digging of the Brussels–Charleroi Canal, which from 1832 onward made waterborne transport possible from as far as the province of Hainaut to the port of Antwerp via the capital.

Independence of Belgium

In 1830 came revolution; Belgium won its independence, and, in the constitution adopted by the National Congress, Brussels, which had played a major role in the uprising against the Dutch, was named the capital of Belgium and the seat of government. As a result, the city's growing political and administrative role enhanced its importance as an economic and financial centre. With a population of more than 123,000 in 1846, it became the central node of Belgium's road and railway network. Its material infrastructure was greatly improved by means of a modern sewage and water supply system (1854–55), the introduction of public transport, and the development of new residential districts. The fragmented local administration, however, was to be streamlined only partly and gradually, as suburban areas were incorporated into the agglomeration; but the influence of the Brussels elite in Belgian national politics remained predominant throughout the 19th century.

THE 20TH CENTURY

In World War I the German occupation of Belgium lasted from August 1914 to November 1918. Numerous social relief movements were instituted; among them, the National Committee for Relief and Food (Comité National de Secours et d'Alimentation) had its headquarters in Brussels and with U.S. aid organized the feeding of the Belgian population. Adolphe de Max, the burgomaster of Brussels, acquired fame for his resistance to the abuses of the German occupiers. The Belgian army reoccupied the capital on Nov. 18, 1918, and four days later King Albert I and Queen Elisabeth reentered the city in state. After the war, the administrative expansion of metropolitan Brussels that had begun in the 19th century continued. The area grew from nine municipalities in 1878 to 16 in 1932 and to 19 in 1954. This metropolitan, or Greater, Brussels became an officially bilingual city in 1932.

In World War II Brussels fell to the invading German army on May 18, 1940. The city did not suffer extensive physical damage but was subjected to harsh terms of occupation. To facilitate control, General Eggert Reeder, chief of the German military administration for Belgium, decided to follow the Nazi policy of creating large urban zones by amalgamating the communes. In order to crush the spirited opposition to this measure, Reeder dissolved all municipal councils and dismissed Joseph van de Meulebroeck, the leader of the opposition, from his post as burgomaster of Brussels. Reeder then appointed a governing council headed by Jan Grauls, a pro-Nazi Flemish nationalist. As in World War I, Germany tried to divide the nation by supporting partisans of Flemish autonomy. Although few Flemings actually collaborated with the enemy, anti-Flemish feelings ran high in metropolitan Brussels. The city was liberated on Sept. 3, 1944, by the British. Five days later the legitimate Belgian government returned to its capital from London.

Since World War II, Brussels has become decidedly more international, with the establishment of the EEC Commission and Council offices in Inner Brussels (1958) and the move of the NATO headquarters from Paris to the northeastern commune of Evere (1967). The city also was host to a successful world's fair in 1958, which helped to rejuvenate the weakened postwar economy.

The Fleming-Walloon conflict

Domestically, however, the situation has been less harmonious. The city has repeatedly been at the centre of tensions between the Flemish and Walloon communities of Belgium, as the Flemings have pressed for effective bilingualism in the public services in Brussels itself and have opposed any further expansion of the mainly French-speaking metropolis into neighbouring Flemish areas. Massive Flemish demonstrations against "Frenchification and territorial annexation" were held in the streets of Brussels in 1961 and 1962. In an attempt to settle the issue, Parliament passed a law in 1963 that restricted the capital to its 19 officially bilingual municipalities but extended language facilities to French-speaking minorities in six suburban boroughs. The Francophone countermobilization against what was regarded as Flemish interference in city affairs led to the formation of the Brussels-based Francophone Democratic Front in 1964. Whereas the Flemings were intent on preventing the Francophone influence from spreading further, the French-speaking residents of Brussels resented the imposition of a legal *carcan,* or "straitjacket," on the city. The front's rapid growth gave it a firm political hold in the late 1970s, but its demise in the early 1980s was equally rapid, as the deepening national and international economic crisis drew attention away from the language conflict. The faltering economy also temporarily halted the reforms begun in 1962 devolving power from the central government to the communities and regions. The exact status of the capital in relation to Flanders and Wallonia thus remained unresolved.

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Bryophytes

Bryophytes (division Bryophyta) are among the simplest of the terrestrial plants. Most representatives lack complex tissue organization, yet they show considerable diversity in form and ecology. They are widely distributed throughout the world and are relatively small compared with most seed-bearing plants. Most are 2–5 centimetres (0.8–2 inches) tall or, if reclining, generally less than 10 centimetres long. The phylum Bryophyta (Figures 1 and 2) includes three main evolutionary lines: the mosses (class Bryopsida, or Musci), the liverworts (class Hepatopsida, or Hepaticae), and the hornworts (class Anthocerotopsida, or Anthocerotae). It is conservatively estimated that there are more than 1,000 genera and more than 18,000 species of bryophytes.

The bryophytes show an alternation of generations between the independent gametophyte generation, which produces the sex organs and sperm and eggs, and the dependent sporophyte generation, which produces the spores. In contrast to vascular plants, the bryophyte sporophyte usually lacks a complex vascular system and produces only one spore-containing organ (sporangium) rather than many. Furthermore, the gametophyte generation of the bryophyte is usually perennial and photosynthetically independent of the sporophyte, which forms an intimate interconnection with the gametophytic tissue, especially at the base, or foot, of the sporophyte. In most vascular plants, however, the gametophyte is dependent on the sporophyte. In bryophytes the long-lived and conspicuous generation is the gametophyte, while in vascular plants it is the sporophyte. Structures resembling stems, roots, and leaves are found on the gametophore of bryophytes, while these structures are found on the sporophytes in the vascular plants. The sporophyte releases spores, from which the gametophytes ultimately develop.

In some bryophytes, sporophytes are unknown. The gametophyte in these bryophytes often reproduces asexually, or vegetatively, by specialized masses of cells (gemmae) that are usually budded off and ultimately give rise to gametophytes. Fragmentation of the gametophyte also results in vegetative reproduction: each living fragment has the potential to grow into a complete gametophyte.

The mature gametophyte of most bryophytes is leafy, but some liverworts and hornworts have a flattened gametophyte, called a thallus (Figure 1). The thallus tends to be ribbonlike in form and is often compressed against the substratum to which it is generally attached by threadlike structures called rhizoids. Rhizoids also influence water and mineral uptake.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 313, and the *Index.*

This article is divided into the following sections:

GENERAL FEATURES

Thallose versus leafy bryophytes

Thallose bryophytes vary in size from a length of 20 centimetres (8 inches) and a breadth of 5 centimetres (2 inches; the liverwort *Monoclea*) to less than 1 millimetre (0.04 inch) in width and less than 1 millimetre in length (male plants of the liverwort *Sphaerocarpos*). The thallus is sometimes one cell layer thick through most of its width (*e.g.,* the liverwort *Metzgeria*) but may be many cell layers thick and have a complex tissue organization (*e.g.,* the liverwort *Marchantia*). Branching of the thallus may be forked, regularly frondlike, digitate, or completely irregular. The margin of the thallus is often smooth but is sometimes toothed; it may be ruffled, flat, or curved inward or downward.

Adapted from W.B. Schofield, *Introduction to Bryology* (1985); McGraw-Hill Book Co., Inc.

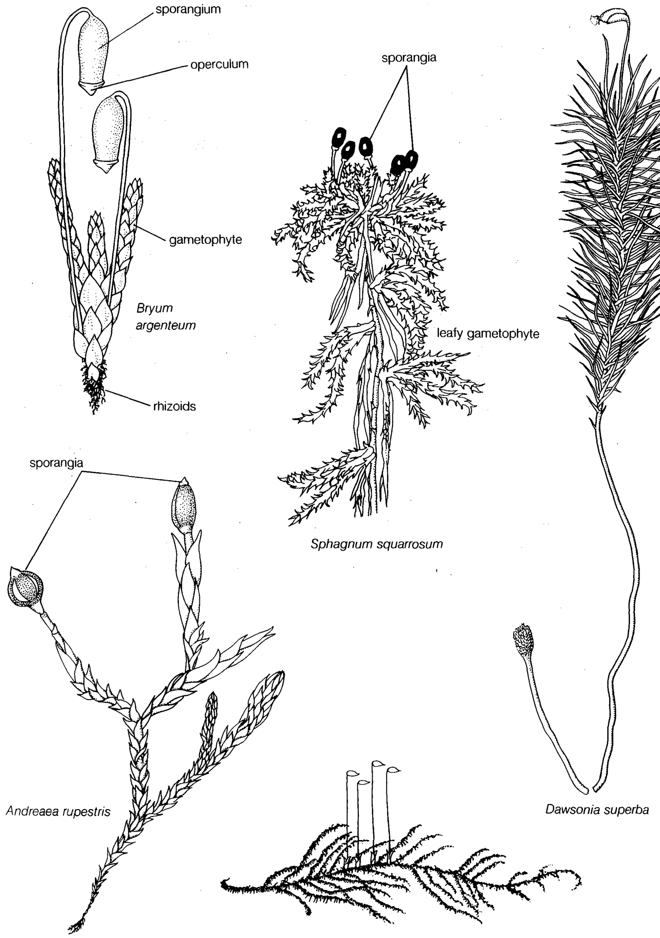

Figure 1: Representative mosses.

Leafy bryophytes grow up to 65 centimetres (2 feet) in height (the moss *Dawsonia*) or, if reclining, reach lengths of more than 1 metre (3.3 feet; the moss *Fontinalis*). They are generally less than 3 to 6 centimetres tall, and reclining forms are usually less than 2 centimetres long. Some, however, are less than 1 millimetre in size (the moss *Ephemerum*). Leaves are arranged in rows of two or three or more around a shoot or may be irregularly arranged (*e.g.,* the liverwort *Takakia*). The leafy shoot may or may not appear flattened. Leaves are usually attached by an expanded base and are mainly one cell thick. Many mosses, however, possess one or more midribs several cells in thickness. Leaves of liverworts are often lobed, while those of mosses are unlobed. Leaves diverge outward from the shoot; rigidity results from water pressure within the cells or from the support of a midrib, when present.

Most gametophytes are green, and all except the gametophyte of the liverwort *Cryptothallus* have chlorophyll.

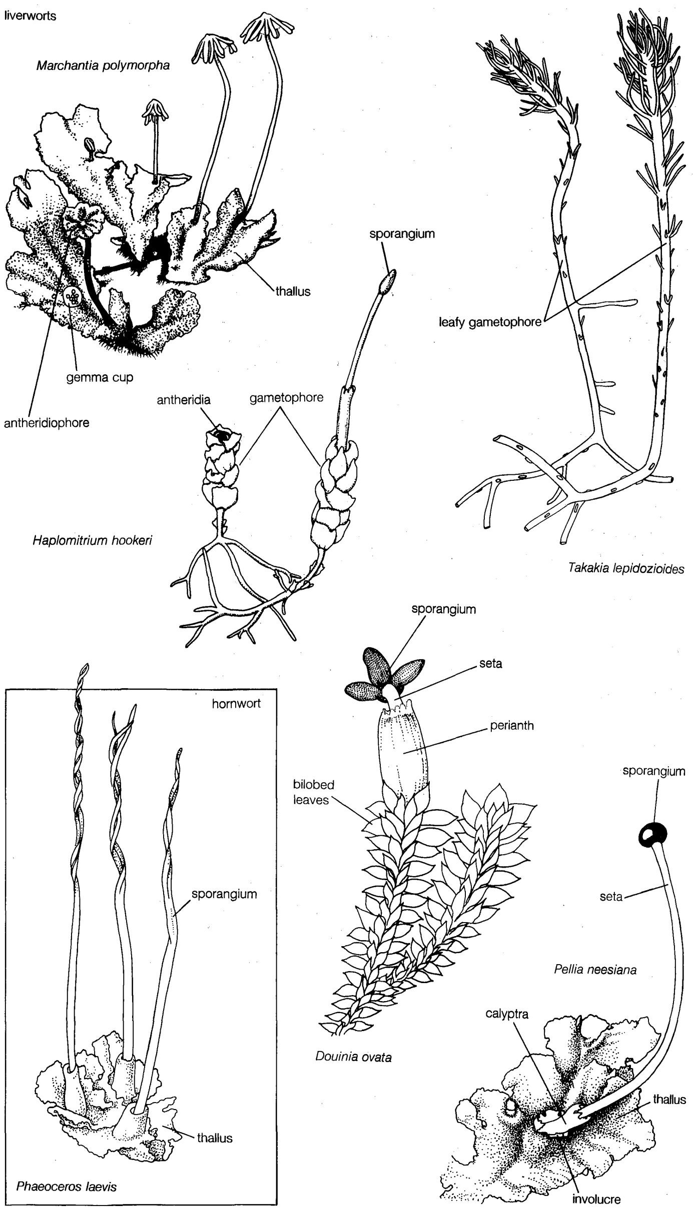

Figure 2: A hornwort and representative liverworts.

Adapted from W.B. Schofield, *Introduction to Bryology* (1985); McGraw-Hill Book Co., Inc.

Many have other pigments, especially in the cellulosic cell walls but sometimes within the cytoplasm of the cells.

Bryophytes form flattened mats, spongy carpets, tufts, turfs, or festooning pendants. These growth forms are usually correlated with the humidity and sunlight available in the habitat.

Importance of water

Distribution and abundance. Bryophytes are distributed throughout the world, from polar and alpine regions to the tropics. Water must, at some point, be present in the habitat in order for the sperm to swim to the egg (see below *Natural history*). Bryophytes do not live in extremely arid sites or in seawater, although some are found in perennially damp environments within arid regions and a few are found on seashores above the intertidal zone. A few bryophytes are aquatic. Bryophytes are most abundant in climates that are constantly humid and equable. The greatest diversity is at tropical and subtropical latitudes. Bryophytes (especially the moss *Sphagnum*) dominate the vegetation of peatland in extensive areas of the cooler parts of the Northern Hemisphere.

The geographic distribution patterns of bryophytes are similar to those of the terrestrial vascular plants, except that there are many genera and families and a few species of bryophytes that are almost cosmopolitan. Indeed, a few species show extremely wide distribution. Some botanists explain these broad distribution patterns on the theory that the bryophytes represent an extremely ancient group of plants, while others suggest that the readily dispersible small gemmae and spores enhance wide distribution.

The distribution of some bryophytes, however, is extremely restricted, yet they possess the same apparent dispersibility and ecological plasticity as do widespread bryophytes. Others show broad interrupted patterns that are represented also in vascular plants.

Peat moss

Importance to humans and ecology. The peat moss genus *Sphagnum* is an economically important bryophyte. The harvesting, processing, and sale of *Sphagnum* peat is a multimillion-dollar industry. Peat is used in horticulture, as an energy source (fuel), and, to a limited extent, in the extraction of organic products, in whiskey production, and as insulation.

Bryophytes are very important in initiating soil formation on barren terrain, in maintaining soil moisture, and in recycling nutrients in forest vegetation. Indeed, discerning the presence of particular bryophytes is useful in assessing the productivity and nutrient status of forest types. Further, through the study of bryophytes, various biological phenomena have been discovered that have had a profound influence on the development of research in such areas as genetics and cytology.

NATURAL HISTORY

Reproduction and life cycle. The life cycle of bryophytes (Figures 3 and 4) consists of two stages, or generations, called the sporophyte and the gametophyte. Each generation has a different physical form. When a spore germinates, it usually produces the protonema, which precedes the appearance of the more elaborately organized gametophytic plant, the gametophore, which produces the sex organs. The protonema is usually threadlike and is highly branched in the mosses but is reduced to only a few cells in most liverworts and hornworts. The protonema stage in liverworts is usually called a sporeling in other bryophytes (see below *Form and function*).

The gametophore—the thallose or leafy stage—is generally perennial and produces the male or female sex organs, or both. The female sex organ is a flask-shaped structure called the archegonium. The archegonium contains a single egg enclosed in a swollen lower portion that is more than one cell thick. The neck of the archegonium is a single cell layer thick and sheathes a single thread

The archegonium and antheridium

From R.R. Ireland, *Moss Flora of the Maritime Provinces* (Ottawa, 1982)

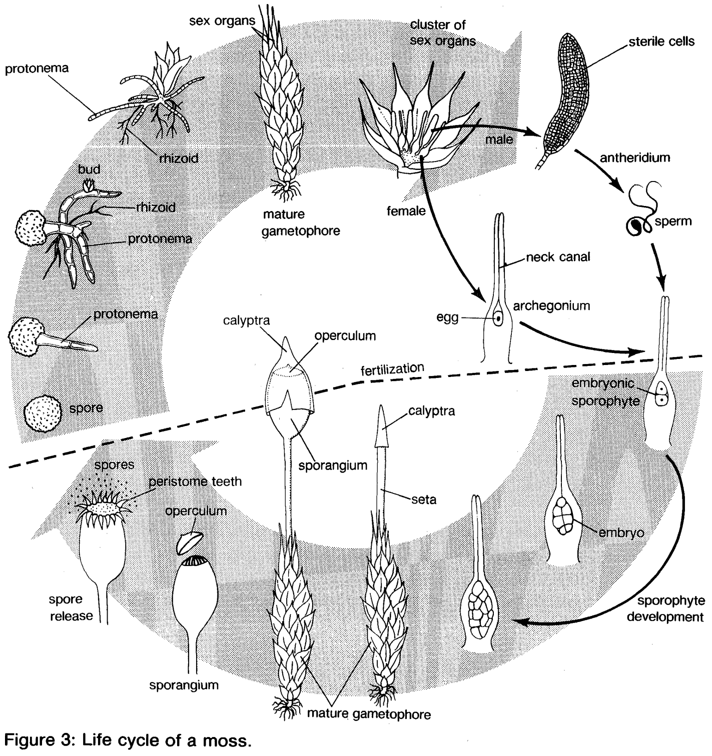

Figure 3: Life cycle of a moss.

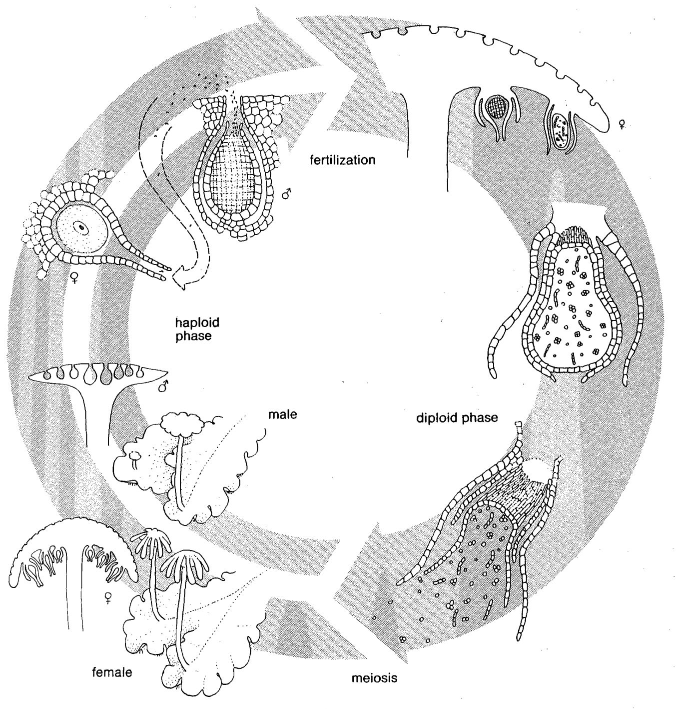

Figure 4: Life cycle of a liverwort (*Marchantia*).

of cells that forms the neck canal. When mature and completely moist, the neck canal cells of the archegonium disintegrate, releasing a column of fluid to the neck canal and the surrounding water. The egg remains in the base of the archegonium, ready for fertilization. The male sex organ, the antheridium, is a saclike structure made up of a jacket of sterile cells one cell thick; it encloses many cells, each of which, when mature, produces one sperm. The antheridium is usually attached to the gametophore by a slender stalk. When wet, the mature antheridium ruptures the jacket to release the sperm into the water. Each sperm has two flagella and swims in a corkscrew pattern. When a sperm enters the field of the fluid diffused from the neck canal, it swims toward the site of greatest concentration of this fluid, therefore down the neck canal to the egg. Upon reaching the egg, the sperm burrows into its wall, and the egg nucleus unites with the sperm nucleus to produce the diploid zygote. The zygote remains in the archegonium and undergoes many mitotic cell divisions to produce an embryonic sporophyte. The lower cells of the archegonium also divide and produce a protective structure, called the calyptra, that sheathes the growing embryo.

As the sporophyte enlarges, it is dependent on the gametophore for water and minerals and, to a large degree, for nutrients manufactured by the gametophore. The water and nutrients enter the developing sporophyte through the tissue at its base, or foot, which remains embedded in the gametophyte. Mature bryophytes have a single sporangium (spore-producing structure) on each sporophyte. The sporangium generally terminates an elongate stalk, or seta, when the sporangium is ready to shed its spores. The sporangium rupture usually involves specialized structures that enhance expulsion of the spores away from the parent gametophore.

Nutrition. Bryophytes generate their nutrient materials through the photosynthetic activity of the chlorophyll pigments in the chloroplasts. In addition, most bryophytes absorb water and dissolved minerals over the surface of the gametophore. Water retention at the surface is assisted by the shape and overlapping of leaves, by an abundance of rhizoids, or by capillary spaces among these structures. Water loss through evaporation is rapid in most bryophytes.

Conducting systems

A few bryophytes possess elaborate internal conducting systems (see below *Form and function*) that transfer water or manufactured nutrients through the gametophore, but most conduction is over the gametophore surface. In most mosses, water and nutrient transfer from the gametophore to the developing sporangium takes place along the seta and also via an internal conducting system. A protective cuticle covers the seta, reducing water loss. The calyptra that covers the developing sporangium prevents water loss in this fragile immature structure. In liverworts the sporangium remains close to the gametophore until it is mature; thus, a conducting system is not formed in the seta. In most hornworts there is also an internal conducting system within the developing horn-shaped sporangium. The internal movement of fluid in all parts of the bryophyte is extremely slow. Storage products include starch and lipids.

Ecology and habitats. Some bryophytes are unusually tolerant of extended periods of dryness and freezing, and, upon the return of moisture, they rapidly resume photosynthesis. The exact mechanism involved remains controversial.

Many bryophytes grow on soil or on the persistent remains of their own growth, as well as on living or decomposing material of other plants. Some grow on bare rock surfaces, and several are aquatic. The main requirements for growth appear to be a relatively stable substratum for attachment, a medium that retains moisture for extended periods, adequate sunlight, favourable temperature, and, for richest luxuriance, a nearly constantly humid atmosphere.

Unusual habitats include decomposing animal waste (many species in the moss family Splachnaceae), somewhat shaded cavern mouths (the liverwort *Cyathodium* and the mosses *Mittenia* and *Schistostega*), leaf surfaces (the moss *Ephemeropsis* and the liverwort genus *Metzgeria* and many species of the liverwort family Lejeuneaceae), salt pans (the liverwort *Carrpos*), bases of quartz pebbles (the moss *Aschisma*), and copper-rich substrata (the moss *Scopelophila*).

In humid temperate or subtropical climates, bryophytes often grow profusely, forming deep, soft carpets on forest floors and over rock surfaces, sheathing trunks and branches of trees and shrubs, and festooning branches. In broad-leaved forests of temperate areas, trees and boulders often harbour rich bryophyte stands, but it is near watercourses that bryophytes tend to reach their richest luxuriance and diversity.

In Arctic and Antarctic regions, bryophytes, especially mosses, form extensive cover, especially in wetlands, near watercourses, and in sites where snowmelt moisture is available for an extended part of the growing season. There they can dominate the vegetation cover and control the vegetation pattern and dynamics of associated plants. The same is true for alpine and subalpine environments in which many of the same species are involved.

Nutrient cycling

Bryophytes, especially mosses, are important in nutrient cycling, in some cases making use of limited precipitation and airborne minerals that are thus made unavailable to the seed plant vegetation. Rapid evaporation from the moss mat is probably critical to some vegetation types by impeding moisture penetration to the root systems of seed plants and therefore indirectly controlling the vegetational composition of some forests.

Bryophytes are fundamental to the development of wetland habitats, especially of peatland. The moss genus *Sphagnum* leads to the development of waterlogged masses of highly acid peatland, in which decomposition is relatively slow. The formation of extensive bogs can control the hydrology of much of the surrounding landscape by behaving like a gigantic sponge that absorbs and holds vast quantities of water and influences the water table. Extension of this saturated living moss mat into living forest can drown the root systems of the forest trees, killing the forest and replacing it with bog. Peatland can also develop on calcareous terrain through the growth of other mosses, including species of the genera *Drepanocladus* and *Calliergon.* These mosses also build up a moss mat that, through organic accumulation of its own partially decomposed remains, alters the acidity of the site and makes it attractive to the formation of *Sphagnum* peatland.

Bryophytes, especially mosses, colonize bare rock surfaces, leading ultimately to the initiation of soil formation. This in turn produces a substratum attractive to seed plant colonists that invade these mossy sites and, through their shading, eliminate the pioneer mosses but create a shaded habitat suitable for other bryophytes. These new colonists, in turn, are important in nutrient cycling in the developing forest vegetation.

FORM AND FUNCTION

The gametophyte form shows several developmental stages: the spore, the protonema, and the gametophore, which produces the sex organs. Spores of bryophytes are generally small, 5–20 micrometres on the average, and usually unicellular, although some spores are multicellular and considerably larger. Spores have chlorophyll when released from the sporangium. They are generally hemispheric, and the surface is often elaborately ornamented.

Spores

The protonema, which grows directly from the germinating spore, is in most mosses an extensive, branched system of multicellular filaments that are rich in chlorophyll. This stage initiates the accumulation of hormones that influence the further growth of newly formed cells. When specific concentrations of the hormones are reached, the branches of the protonema generate small buds, which in turn produce the leafy gametophore.

In most liverworts and hornworts, the protonema is usually limited to a short unbranched filament that rapidly initiates a three-dimensional cell mass, the sporeling. This sporeling is rich in chlorophyll and soon forms an apical cell from which the gametophore grows.

In moss gametophores the leaves of the shoots are spirally arranged on the stem in more than three rows. Leaves often have elaborate ornamentation on the cell surfaces. This ornamentation is often important in rapid water uptake. Although the leaf begins its growth from an apical cell, cells are soon cut off between the apical cell and the leaf base, and further division of these cells results in the elongation of the leaf and also in the production of one or more midribs. The gametophore is often attached to the substratum by rhizoids. The rhizoids are structurally similar to cells of the protonema, but they lack chlorophyll. In some mosses, rhizoids closely invest the stem among the leaf bases and perform a significant function in external water conduction and retention before its absorption by stem and leaves.

The internal structure of the stems of moss gametophores is usually simple. The outer cells are often thick-walled and supportive, while the inner cells are generally larger and have thinner walls. Some mosses, however, have considerable tissue differentiation in the stem. In the moss subclass Polytrichidae, for example, a complex conducting strand is often formed in the centre of the stem. It consists of an internal cylinder of water-conducting cells (the hydroids) surrounded by layers of living cells (leptoids) that conduct the sugars and other organic substances manufactured by the gametophore. This conducting system is analogous to that of the vascular plants, except that it lacks lignin (a carbohydrate polymer), and it closely resembles that found in the fossils of the earliest land plants.

In gametophores of leafy liverworts, the leaves are arranged in two or, usually, three rows. The plants are often flattened horizontal to the substratum. Lobing of these leaves is sometimes complex, as is their orientation on the stems as compared with the mosses. Rhizoids are generally confined to the undersurface of the stem and are important in that they form attachments and influence water retention and uptake by the leafy plant.

In gametophores of thallose liverworts and hornworts, an internal conducting strand is rarely developed. In a few genera of the liverwort order Metzgeriales, the water-conducting cells have a form similar to water-conducting cells of vascular plants, but the cells of the liverworts and hornworts, like those of mosses, lack the lignin that characterizes the cell walls of water-conducting cells of vascular plants.

The thalli of most liverworts and hornworts consist of relatively undifferentiated layers of cells. Those cells on the dorsal surface are rich in chlorophyll, while those situated deeper within the thallus lack chlorophyll but have storage products of photosynthesis, especially starch. Fungi are often present in the cells of many thalli (and also leafy liverwort stems) and are probably important in water and mineral uptake as well as in making organic compounds available for the nutrition of the gametophore. The thalli of the liverwort order Marchantiales show considerable tissue differentiation, which gives these complex thalli a structure analogous to that of the leaves of vascular plants and provides structural features which allow them to occupy habitats too dry for many other liverworts and hornworts.

Thalli

The sexual reproduction of bryophyte gametophores is usually seasonally restricted, often initiated by short-day or long-day illumination; thus, especially in temperate climates, sex organs appear and mature in the autumn, while in more extreme climates they appear in the spring or summer. In mosses, the sex organs are usually sheathed by specialized leaves and are embedded in a mass of filaments that protects the sex organs from drying out before maturity. Many mosses have antheridia and archegonia on separate gametophores, ensuring outbreeding, while others have both sexes on the same gametophore but apparently with features that discourage inbreeding.

The perianth

In many leafy liverworts the archegonia are often enclosed by a protective sleeve, the perianth, and have mucilage hairs among them with a function similar to that of the paraphyses of mosses. The antheridia of leafy liverworts are often on specialized branches and at the axils of specialized leaves that are usually swollen to enclose them. Most leafy liverworts have antheridia and archegonia on separate plants.

The archegonia of the hornworts are completely embedded in the dorsal surface of the thallus, while antheridia are found in chambers near the dorsal surface. Thalli may contain antheridia or archegonia or both.

Sporophytes of mosses usually consist of the foot, which penetrates the gametophore, the seta, with an internal conducting system, and a terminal sporangium. The seta contains chlorophyll when immature and cannot absorb moisture from the environment because its surface is covered by a water-impermeable layer, the cuticle. The sporophyte is photosynthetic when immature, but its restricted amount of chlorophyll-containing tissue rarely produces enough carbohydrates to nourish a developing sporangium. All water and much of the needed nutrients are absorbed from the gametophore and are conducted through the transfer tissue of the foot up the conducting strand that leads to the apex of the sporophyte. The seta is made rigid by thick-walled cells external to the conducting strand. The sporangium differentiates after the seta elongates and is protected from injury and drying by the calyptra.

The moss sporangium usually opens by way of an apical lid (the operculum). When the operculum falls, there is exposed a ring of teeth that controls the release of the spores over an extended period of time. These teeth usually respond to slight moisture changes and pulsate inward and outward, carrying spores out of the sporangium on their jagged inner surfaces. In the moss subclass Polytrichidae, however, the tiny spores exit through a series of holes between the teeth and a membrane that closes much of the mouth; thus, any slight movement of the sporangium causes spores to shake out into the air. In the moss subclass Andreaeidae, the spores are released when the sporangium wall gapes open in longitudinal slits. In the genus *Sphagnum,* air is trapped within the sporangium as it matures; as the sporangium dries out, it shrinks, until the buildup of internal pressure abruptly shoots the operculum and spores into the air.

In most liverworts, the sporangium matures before the seta elongates, pushing the sporangium above the calyptra that protected it. Elongation is rapid, and the seta is held erect by water pressure within its cells. The sporangium usually contains within it elongate cells (elaters) with coiled thickenings that are scattered among the spores. When the sporangium opens, usually very rapidly when dry, it does so along four longitudinal lines, exposing the elaters, which uncoil rapidly and throw themselves and the adjacent spores into the air. Other devices exist for spore release in the liverworts.

Hornworts are unusual among the bryophytes because the sporophyte has indeterminate growth. This means that throughout the growing season new tissue is continually produced, even when spores are being shed. Early in its growth within the archegonium, the embryo produces a foot that penetrates the thallus and an apical meristem that elongates the rest of the horn-shaped sporophyte to rupture the thallus surface. A meristem (an area of actively dividing cells that gives rise to all subsequent tissue) is soon differentiated just above the foot, between it and the horn-shaped sporophyte above, and this meristem contributes new growth to the elongating sporophyte throughout the growing season and ceases when the gametophore disintegrates around it. The sporophyte thus matures near the apex while new tissue is differentiated just above the foot, contributing to the elongation of the sporophyte. The sporangium usually opens by two longitudinal lines on opposite sides of the horn. As the apex matures, it exposes the spores and elaters, which are released to the air.

EVOLUTION AND PALEONTOLOGY

Poor fossil record

The fossil record of bryophytes is poor. Some fossils, however, show a morphology, size, and cellular detail that characterize bryophytes, and the specimens are treated as fossil bryophytes. Since sex organs and attached sporophytes are absent in nearly all fossil material and because the gametophytes of some living vascular plants resemble the gametophores of some bryophytes, the assignment of these fossils as bryophytes is by no means secure.

The earliest presumed bryophyte fossils are contemporaneous with the earliest vascular plants of the Late Devonian epoch, 374 to 360 million years ago. These fossils structurally resemble gametophores of the liverwort order Metzgeriales. Indeed, fossil material of the Carboniferous period (360 to 286 million years ago) also is structurally similar to genera of Metzgeriales. The specimens are surprisingly well preserved and show considerable cellular detail.

The most elegantly preserved bryophyte fossils are those in amber of the Eocene epoch (57.8 to 36.6 million years ago). The detailed cellular structure and morphology of the gametophore make the determination of the genus reasonably secure. The genera are still extant, although not where the fossil material was found, and even the species relationships can be suggested.

For mosses, the earliest material that appears unambiguous is in the Permian period (286 to 245 million years ago), and the detailed relationships are not clear. The subclass Bryidae is most likely, but more precise attribution is difficult.

Well-preserved material of mosses and liverworts appears in the Tertiary period (66.4 to 1.6 million years ago), and most of the main evolutionary lines are represented. Fossils of the Neogene (roughly equivalent to the last one-third of the Tertiary period, 23.7 to 1.6 million years ago) are relatively numerous, and subfossil material of the Quaternary period (1.6 million years ago to the present) can be determined with confidence as modern species. Mosses are most richly represented in this material, and species of wetland habitats predominate in the record.

CLASSIFICATION

Classification of the liverworts leans heavily on gametophyte structure, with sporophyte structure providing additional evidence of relationships. In the hornworts and mosses, the structure of the sporophyte, especially the sporangium, is important in distinguishing the main evolutionary lines, while gametophytic features provide the details for distinguishing genera and species.

Annotated classification. The classification presented here reflects main evolutionary lines. These seem best illustrated at the order level. Those orders that are considered to be most generalized are treated first; and those most specialized, last.

CLASS HEPATOPSIDA (or HEPATICAE; liverworts)

Protonema generally reduced to a few cells, with gametophore differentiated early after spore germination; rhizoids unicellular; gametophore leafy or thallose and generally flattened; sex organs lacking paraphyses; leaves lacking true midrib; leaf cells often with corner thickenings; complex oil bodies often in cells of gametophore; sporangium jacket lacking stomata, and often with transverse thickenings in cellwalls; sporangium usually opening by longitudinal lines; sporangium releasing all spores and elaters at the time it opens; calyptra remaining at base when seta elongates.

Order Takakiales

Leaves cylindrical and irregularly but radially arranged on an erect shoot that arises from a subterranean, colourless, rootlike system; sex organs lateral but near shoot apex; rhizoids absent; sporophytes unknown; Southeast Asian and northwestern American distribution; a single genus, *Takakia,* with two species.

Order Calobryales

Leaves flattened and in three rows on an erect shoot arising from a colourless, subterranean, rootlike system that lacks rhizoids; sex organs lateral but near shoot apices; sporophytes with elongate seta; sporangium elongate, with elaters and thickenings on the jacket cell walls; opening by 1–4 longitudinal lines; mainly of mid-latitudes, most species in the Australasian and Indo-Malayan region; 2 genera, *Haplomitrium* (12 species) and *Steereomitrium* (1 species).

Order Metzgeriales

Thallose, with the thallus mainly of uniformly thickened cell walls, usually reclining but sometimes erect; branching varies from forked to regularly pinnate or irregular; smooth rhizoids on the undersurface; sex organs lateral; sporophytes with elongate seta; sporangia spherical to elongate, with elaters and thickenings of the jacket cell walls; opening by 1–4 longitudinal lines or irregularly; widely distributed throughout the world; approximately 30 genera and 550 species; sometimes the order Treubiales is separated from this order.

Order Jungermanniales

Leaves flattened, in 2 or 3 rows, usually broadened to attachment, often lobed; shoots reclining, erect, or pendent; rhizoids smooth-walled; archegonia terminating shoot, surrounded by a chlorophyllose sheath (perianth); sporophyte with seta; sporangium spherical to elongate, with elaters and thickenings of the jacket cell walls, opening by 4 longitudinal lines (rarely helical); distributed throughout the world, reaching greatest abundance in humid subtropical to temperate climates; contains at least 85 percent of the liverworts; conservatively, 300 genera and more than 7,000 species.

Order Sphaerocarpales

Essentially lobate thallus in all modern representatives; thallus of parenchyma cells reclining or erect, with smooth-walled rhizoids; each sex organ surrounded by an enveloping sac, lateral; sporangium spherical, lacking seta and elaters, opening by disintegration of the unornamented jacket cells; terrestrial except the aquatic genus *Riella;* distributed mainly in milder temperate climates; 3 genera with approximately 30 species.

Order Monocleales

Large thalli of mainly uniformly parenchymatous cells, reclining; thallus forked to irregularly branched; archegonia within a sleevelike chamber behind the lobe apex; antheridia in padlike receptacles in the same location on different thalli; sporangia elongate on a massive elongate seta, with long elaters and opening by a single longitudinal line; jacket with thickenings on cell walls; in South and Central America and New Zealand; a single genus, *Monoclea,* with 1 species.

Order Marchantiales

Thallus often of complex anatomy, with air pores on the dorsal surface, air chambers with chlorophyllose cells forming a photosynthetic area, and cells of the remainder of the thallus serving for storage; ventral scales often present; rhizoids; sex organs sometimes borne on a stalked receptacle; sporophytes with short seta or seta absent; sporangia spherical or elongate, opening by regular or irregular longitudinal lines, a caplike lid, or decomposition; sporophytes often carried up from the thallus surface by elongation of the stalk of a receptacle, with the sporangia hanging downward; occupying a diversity of habitats—some can withstand extended periods of dryness while others are floating or submerged aquatics, and still others grow in humid shaded sites—approximately 27 genera and 450 species widely distributed throughout the world; the genus *Riccia* containing nearly half the species of the order.

CLASS ANTHOCEROTOPSIDA (or ANTHOCEROTAE; hornworts)

Protonema reduced to short filament or absent, differentiating the gametophore early after spore germination; rhizoids unicellular and smooth-walled; gametophore thallose, sometimes lobate; archegonium not a discrete structure, made up of an egg and neck canal cells embedded in the dorsal surface of the thallus; often several antheridia within a chamber embedded in the dorsal surface of the thallus; thallus sometimes with ventral pores, sometimes developing mucilage chambers; thallus lacking complex oil bodies; chloroplasts often solitary in each cell and often with pyrenoid; sporangium horn-shaped, usually with stomata in jacket; elaters often multicellular and often lacking helical thickenings; columella of sterile tissue extending the length of the sporangium, with the spore-bearing tissue overarching and sheathing it; sporangium indeterminate in growth from a basal meristem just above the foot; spores shed throughout the growing season by longitudinal lines of openings extending from the apex downward as the sporangium ages, sometimes (in *Notothylas*) by decomposition of the sporangium jacket.

Order Anthocerotales

Characteristics are those of the class; widely distributed in temperate to tropical latitudes, with greatest diversity in the tropics and subtropics; containing 6 or 7 genera and probably fewer than 300 species.

CLASS BRYOPSIDA (or MUSCI; mosses)

Protonema an extensive many-branched filament that precedes gametophore production; rhizoids multicellular, branched; gametophore leafy, with leaves spirally arranged, usually in more than 3 rows; gametophore usually not strongly flattened; sex organs usually with paraphyses among them; leaves unlobed and often with thickened midrib; cells usually lacking corner thickenings; oil bodies, if present, not complex; jacket of sporangium often with stomata; sporangium usually opening by apical cap (operculum); peristome teeth usually surrounding the sporangium mouth and influencing spore release; columella usually present, encircled or overarched by a spore-bearing layer; calyptra capping apex of elongating seta and influencing survival and differentiation of sporangium; spores generally shed over extended period; seta a rigid structure with internal conducting strand and holding sporangium well above gametophore in most instances.

Subclass Andreaeidae

Sporophytes usually lacking a seta; sporangium opening by longitudinal lines; sporangium with spore-bearing layer overarching and encircling the central columella; gametophore irregularly branched, dark-pigmented, with spirally arranged leaves, attached to the substratum by rhizoids; leaves with or without midrib; paraphyses few or absent; sporophytes usually pushed beyond perichaetium on an elongate leafless extension of the gametophore (pseudopodium); mainly in cooler climates throughout the world, confined mainly to siliceous rock surfaces; 2 genera, *Andreaea* and *Andreaeobryum*, with probably fewer than 100 species.

Subclass Sphagnidae

Sporophytes lacking a seta; subspherical sporangium opening by a lid (operculum) released explosively with the spores when sporangium dries, shrinks in diameter, and reaches high atmospheric pressure through compression of the gases within; protonema phase thalloid; branching in fascicles; leaf without midrib; leaf cells forming a network of elongate chlorophyllose cells surrounding dead swollen cells reinforced by fibril thickenings in walls and perforated by pores; sporophytes pushed beyond perichaetium by leafless extension of gametophore (pseudopodium); widely distributed in the world but forming extensive peatland mainly in boreal regions; 1 genus, *Sphagnum*, with more than 160 species.

Subclass Tetraphidae

Sporophytes with elongate seta; sporangium opening by an operculum exposing four multicellular peristome teeth that respond to moisture change to release spores gradually; spore layer forming a cylinder around central columella; protonema filamentous but with thallose flaps; gametophores erect, with rhizoids at base, leaves with midrib, all cells with chlorophyll; widely distributed in the Northern Hemisphere, with *Tetrodontium* also present, but rare, in the Southern Hemisphere; 2 genera, *Tetraphis* and *Tetrodontium*, with 3 or 5 species.

Subclass Polytrichidae

Sporophytes with elongate rigid seta containing conducting system; sporangium opening by operculum; numerous multicellular peristome teeth in a single concentric circle and overarching a membrane formed by the expanded apex of the columella (many rows of teeth and no membrane in *Dawsonia*); spores very small and released gradually through spaces between the teeth; spore layer forming a cylinder around central columella; gametophores erect, often with complex internal conducting system in stems and often leaves; leaves with numerous chlorophyllose elongate flaps on upper face; widely distributed throughout the world at most latitudes and altitudes, mainly terrestrial; approximately 16 genera and 370 species.

Subclass Buxbaumiidae

Sporophyte with elongate or short seta; sporangium asymmetrical, with operculum; peristome teeth sometimes in several concentric circles, the outer articulated, the inner forming a cone opened at the tip; spores released slowly when slight pressure on the sporangium surface causes the spores to puff out through the narrow mouth; gametophore sometimes extremely reduced and microscopic, always small but sometimes with leaves; widely but erratically distributed in temperate to tropical regions; 4 genera with approximately 40 species.

Subclass Bryidae

Sporophyte may have elongate seta, with or without conducting strand; sporangium diverse in form, with internal cylindric columella encircled by spore-bearing layer, usually opening by operculum to expose articulated peristome teeth in 1 or 2 concentric circles; peristome teeth pulsating in response to moisture changes, extracting the spores from the sporangium and gradually releasing them; gametophores diverse in form and structure; widely distributed throughout the world in most habitats except the sea, representing more than 95 percent of the mosses; more than 650 genera and more than 9,000 species.

Subclass Archidiidae

Sporophyte with no seta; sporangia containing a restricted number of large spores (sometimes 4), lacking columella, opening by decomposition of the jacket; gametophore small, leaves with midrib; attached to substratum by rhizoids; of scattered distribution in temperate to subtropical climates; a single genus, *Archidium*, with approximately 26 species.

Critical appraisal. The order Takakiales is highly controversial; some researchers consider it sufficiently distinctive to separate it from the bryophytes to another phylum, while others merge it with the Calobryales. Discovery of sporophytes could settle the question. The order Anthocerotales is considered by some researchers to be so unrelated to bryophytes that it is placed in its own phylum, the Anthocerotophyta. The evolutionary lines of the class Bryopsida are most easily demonstrated by the subclasses. The treatment of orders and families remains in a state of flux, with widely varying opinions derived from differing interpretations of the taxonomic importance of characteristics. Even phylogenetic placement of the sequence of subclasses is difficult.

Fundamental classification of bryophytes is hampered by a lack of agreement concerning not only the critical features that define a bryophyte but also the criteria that can be used to interpret relationships. Consequently, there are considerable differences among classification systems. It is vital that there be an adequate assessment concerning the diversity of bryophytes that now exist. The extremely limited number of researchers in the field of bryology greatly curtails the acquisition of this information, much of which is being lost when vegetation is destroyed before the floristic structure has been documented and conserved.

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(W.B.S.)

Budapest

Once called the "Queen of the Danube," Budapest, the capital of Hungary, has long been the focal point of the nation and a lively cultural centre. The city straddles the Danube (Hungarian: Duna) River in the magnificent natural setting where the hills of western Hungary meet the plains stretching to the east and south. It consists of two parts, Buda and Pest, which are situated on opposite sides of the river and connected by a series of bridges.

Although the city's roots date to Roman times and even earlier, modern Budapest is essentially an outgrowth of the 19th-century empire of Austria-Hungary, when Hungary was three times larger than the present country. Hungary's reduction in size following World War I did not prevent Budapest from becoming, after Berlin, the second largest city in central Europe. One out of five Hungarians now lives in the capital, which, as the seat of government and the centre of Hungarian transport and industry, dominates all aspects of national life. Tens of thousands of commuters converge on Budapest daily, more than half of the nation's university students attend school in the city, and about half of the country's income from foreign tourism is earned there.

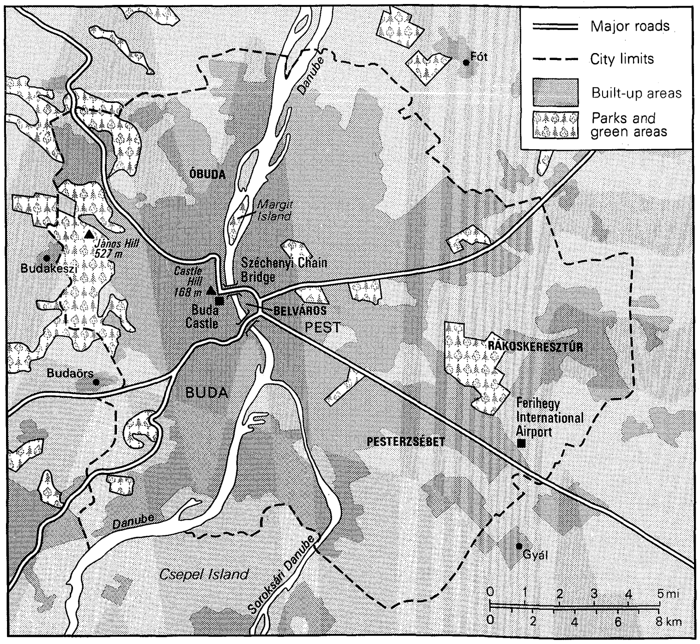
Although Hungary became a member of the Soviet bloc after World War II, the city has long received investment capital from the West, private enterprise has been encouraged, and the city's large black market has been tolerated. The impression of plenty—smart shops, restaurants with high culinary standards, and a casino—has increasingly set Budapest apart from the relatively drab capitals of most eastern European countries.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

The city site. Strategically placed at the centre of the Carpathian Basin, Budapest lies on an ancient route linking the hills of Transdanubia (Hungarian: Dunántúl) with the Great Alfold (Great Hungarian Plain; Hungarian: Nagy Magyar Alföld). The wide Danube was always fordable at this point because of a few islands in the middle of the river. The city has marked topographical contrasts: Buda is built on the higher river terraces and hills of the western side, while the considerably larger Pest spreads out on a flat and featureless sand plain on the river's opposite bank.



Budapest and surrounding area.

Climate. The climate of Budapest is transitional between the extreme conditions of the Great Alfold and the more temperate climate of Transdanubia, with its abundant rainfall. Mean annual temperature is 52° F (11° C), ranging from a July average of 72° F (22° C) to 30° F (−1° C) in January. Mean annual precipitation is 24 inches (600 millimetres). Winter snowfalls can be heavy, and the temperature may fall below 5° F (−15° C), but, on the other hand, heat waves combined with humidity in the summer can make the air oppressive. Flooding in Pest was endemic before the river was regulated in the 19th century. The Danube (blue only in the Johann Strauss waltz) has become heavily contaminated, and air pollution, from which the inhabitants of Buda have largely been able to escape, has afflicted most districts in Pest.

The city layout. *Buda.* Buda was the kernel of settlement in the Middle Ages, and the cobbled streets and Gothic houses of the castle town have preserved its old layout. Until the late 18th century, Pest remained a tiny enclave, but then its population exploded, leaving Buda far behind. In the latter half of the 20th century, growth has been more evenly distributed between the two parts. Contemporary Budapest covers 203 square miles (525 square kilometres), of which about half is built up. Buda's hilltops, still crowned by trees, the Danube flanked by three lower hills, the bridges, Margit (Margaret) Island, and the riverfront of Pest lend a remarkable visual identity to the city.

The hills of Buda

In a central position is Castle Hill (Várhegy), 551 feet (168 metres) above sea level and crowned by the restored Buda Castle (Budai vár, commonly called the Royal Palace). In the 13th century a fortress was built on the site and was replaced by a large Baroque palace during the reign (1740–80) of Maria Theresa as queen of Hungary. The structure was destroyed or damaged and rebuilt several times over the centuries, most recently when it was razed during World War II. Restoration of the palace was completed in the mid-1970s, and it now houses the National Széchenyi Library, the Historical Museum of Budapest (commonly called the Vármúzeum, or Castle Museum), and the Hungarian National Gallery. The spire of the medieval Gothic Church of Our Blessed Lady (Nagyboldogasszony), com-

Budapest from the Fishermen's Bastion on Castle Hill.

Peggy Kahana—Peter Arnold, Inc.

monly called Matthias Church (Mátyás templom), also thrusts into the skyline above Castle Hill, with the late 19th-century Fishermen's Bastion (Halászbástya) in front and a contemporary hotel next to it.

To the south of Castle Hill rises the higher Gellért Hill (771 feet), a steep limestone escarpment overlooking the Danube, which provides a panoramic view of the whole city. At the top stands the Citadel (Citadella)—built by the Austrian army in the mid-19th century in order to keep watch over the town—which serves today as a hotel and restaurant and doubles on St. Stephen's Day (August 20) as the stage for a splendid fireworks display. The Liberation Statue near the Citadel commemorates the victory of the Soviet army over German forces in 1945. Rózsa (Rose) Hill, the third hill near the river, lies north of Castle Hill. It is the most fashionable district of Budapest, where Hungary's elite have houses. The Lukács (Lucas) Bath at the foot of the hill is frequented by Budapest's literati.

Below the three hills stretches the city. Opposite Rózsa Hill lies Margit Island, a mile-long park with hotels and swimming pools. Facing Margit Island on the Pest side is the headquarters of the Hungarian Socialist Party (until 1989 the Hungarian Socialist Workers' Party), an unprepossessing modern block standing next to the ornate Parliament Building (Országház). Designed in Neo-Gothic style and influenced by the Houses of Parliament in London, the building (completed in 1902) has been little used since the end of World War II. Further along the bank lies the Neo-Renaissance building of the Hungarian Academy of Sciences (1862–64); the Vigadó, a large hall built in the Romantic style (1859–64); and two obtrusive modern hotels.

The Belváros

Pest. The heart of Pest is the Belváros (Inner Town), an irregular pentagon with its longest side running parallel to the Danube; only traces of the original town walls remain. The district accommodates offices, parts of the Loránd Eötvös University, and shops. The Váci utca, a narrow street turned pedestrian thoroughfare, is the most fashionable shopping centre of Budapest. The Town Hall (Fővárosi Tanács), a Baroque building erected between 1724 and 1747, is in the northeast corner of the Belváros next to Pest County Hall (Pest megyei Tanács). The Inner Town Parish Church (Belvárosi plébániatemplom) is the oldest building in Pest. Rebuilt in the Baroque style in the 18th century, as were many other churches in Pest and Buda, the church had been the most impressive of medieval Pest. St. Stephen's Crown, the symbol of Hungarian nationhood, is on display in the Hungarian National Museum, a Neoclassical edifice located just outside the Belváros.

Pest's main boulevards form concentric semicircles around the Belváros. The nearest to the centre follows the line of the former city walls. The Nagykörút (Great Boulevard), which formerly bore the names of Habsburg archdukes and archduchesses and is now divided into four named sections, follows the riverbed of a dry tributary of the Danube. Most of the ministries and other government offices are to the north of the Belváros. The Neo-Romanesque St. Stephen's Basilica, the Italian Neo-Renaissance State Opera House, the National Theatre, and the concert hall of the Franz Liszt Academy of Music were all built in the 19th century. The stucco buildings of this eclectic architectural collection have not weathered well.

The finest thoroughfare in Budapest, Népköztársaság útja (People's Democracy Avenue, formerly Andrássy út), runs in a straight line from the centre of Pest to City Park (Városliget), which contains the Millennium Monument. The monument consists of a semicircular pillared colonnade displaying statues of Hungarian kings and national leaders, with a statue of the archangel Gabriel surmounting a 118-foot-high central column. The Museum of Fine Arts, other museums, the Budapest Zoo, the renowned city circus, and an amusement ground (which was once called the "English Park") are all found nearby.

THE PEOPLE

The capital is more than 10 times larger than Hungary's next largest city. The rise of population has been phenomenal: its rate of increase from about 100,000 in the 1840s to 1,000,000 in 1918, for example, far outstripped that of London during the same period. Natural population growth has never been a factor in this expansion. Rather, more die in the city than are born there, the result of a never-ending migration of people from villages and towns to the capital. By the late 20th century, however, the rate of growth had slowed, and the population had begun to shift from the central districts to the periphery. Residential districts—such as Pesterzsébet (Pestszenterzsébet) and Kelenföld in the south, Rákoskeresztúr in the east, and Óbuda, Békásmegyer, and Újpalota in the north—have been growing as the inner city has been redeveloped.

Always a city of marked social divisions, Budapest once revealed the deep-rooted contrast in life-styles between the aristocrats, who built palaces in the town centre, and those who lived in the slum districts and sprawling temporary barrack settlements on the city's perimeter. During the Stalinist period after World War II, these contrasts largely disappeared, but an acute housing shortage has persisted. The majority of the inhabitants continue to live in relatively small flats. Economic reforms since the late 1960s have created new wealth, which, in turn, has sharpened the differences between the more ostentatious life-style of the new middle classes—whose privileged members are able to build second homes in the Buda Hills and on Lake Balaton—and that of the workers who populate the gigantic, faceless housing estates of the drab outlying residential districts.

Ethnic diversity

The capital, apart from its large number of foreign visitors, is entirely Hungarian-speaking; in the past it never was. The remarkably diverse ethnic background of Budapest's population has been one of the city's greatest strengths. In the past Buda was run by German and later German and Hungarian burghers. In the early 19th century the government of Pest was in the hands of German burghers, shipping was controlled by Serbs, and the merchants were largely "Greek" (*i.e.,* Greek and other Balkan peoples). German-speaking industrial workers were brought in from the west and large numbers of Jews moved in from the east. By 1900 nearly one-fourth of the inhabitants of Budapest were Jewish, but the Jewish community was largely destroyed during World War II. Postwar Budapest has become culturally homogeneous.

THE ECONOMY

Developed from early times to be the true centre of the country's economy, the town was originally self-sufficient in food production. Buda's wine was renowned before *Phylloxera* and urban expansion destroyed the vineyards that once covered the higher river terraces of the Danube. Today the rest of Hungary supplies its capital with food, to which the orchards, gardens and nurseries, and dairy farms of the Buda Hills contribute only a small fraction.

Commerce and industry. In the past, all roads converged on Buda, and its market was always famous. By the 19th century, however, commerce had shifted to Pest; its wealthy merchants dominated Hungary's agricultural market, and its banks (by 1900 more than 100) controlled 60 percent of the country's bank capital. Commerce is still concentrated in Pest's Belváros, including the major banks, most of the foreign trading companies, the state travel agency (Idegenforgalmi, Beszerzési, Utazási és Szállítási R.T., or IBUSZ), and the best shops. To help mitigate the growing imbalance in consumer services, large shopping centres have been built in Óbuda, Lágymányos, Újpest, Zugló, and elsewhere. The national economic reforms introduced since 1968 have created a private sector in retailing and in service industries. Both the public and the private sectors have benefited from the increasingly important tourist industry.

Economic dominance

In the last quarter of the 19th century, the city—with neither local raw materials of any sort nor even skilled workers (they had to be imported from Austria and Moravia)—was transformed from a commercial base to the country's most prominent industrial centre. Except for a few engineering factories, manufacturing was at first limited to the processing of raw materials, particularly food, and huge grain mills were built on the Danube. Primary-metal and engineering works (especially for agricultural machinery and ships) and munitions and electronics factories soon followed. Automobiles have been produced since 1905, but light industries grew fast only after World War I. Possessing the bulk of the country's industrial production and, by 1938, about three-fifths of its factory workers, Budapest strangled the economic growth of all the other towns. The city's location on the Danube and especially the construction of the rigidly centralized Hungarian railways, which had preceded the growth of industry, were instrumental in this dominance. The influx of foreign capital, channeled through the large banks and through the government, and the control of large factories by leading Budapest banks reinforced the centralizing role played by the transport system. The loss of two-thirds of Hungary's territory in 1918 and the establishment of a centrally planned economy after 1945 have exacerbated the hypertrophy of Budapest in relation to the rest of the country, although by 1969 the capital's share of the country's industrial work force had been reduced to about one-half. More than half of the factory workers in Budapest are employed in heavy industry; the rest are divided between light industry and food processing.

Transportation. Transportation has been the key to Budapest's rapid expansion. A famous crossing point on the Danube where highways have always converged in the past, it has become the hub of the country's trunk roads and main railway lines, all of which radiate from the capital. It has also developed Hungary's largest bus terminal as well as its only commercial airport, Ferihegy International Airport. Csepel Free Port, downstream from the city centre on Csepel Island, handles international freight cargo on the Danube and is equipped to handle container traffic. The head office of the International Danube Commission is in Budapest. Of the capital's eight bridges, the oldest and best known is the Széchenyi Chain Bridge (Széchenyi Lánchíd), built in the 1840s and named for the 19th-century Hungarian reformer István Széchenyi.

Public transportation

Horse-drawn trams were introduced in 1866; and, after the first electric tramways were built in 1887, they soon expanded into a large network. The trams and buses (the latter introduced in 1915) met local transport needs before the 1940s, but they could not cope with the rapid growth of traffic after World War II. Yet, despite the congestion caused by the colossal expansion in the use of the private automobile and the reduction in the number of tramcars, city transport has once more become efficient. The Metro, a subway system constructed in the 1970s, operates three lines, including one that runs under the river and connects Buda to Pest; it is a showpiece that is clean, fast, and (like the city's buses and old-fashioned rattling trams) absurdly cheap.

The hills of Buda, with their pleasant wooded paths, can be reached easily from the town either by an old cog railway, bus, or a chairlift that takes sightseers to the top of János Hill, which at 1,729 feet (527 metres) above sea level is the highest point in Budapest. The Pioneers' Railway (Úttörővasút), which winds through the hills, is managed largely by children.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Budapest is the seat of the Hungarian government and of the central offices of the Hungarian Socialist Party; it is also the seat of Pest *megye* (county) and Buda *járás* (district). The city is divided into 22 administrative districts—six on the Buda side, 15 in Pest, and Csepel Island—each of which elects a councillor (*tanácstag*). Control of the city is formally vested in the Budapest Metropolitan Council (elected by the district councillors who in turn are elected by the whole adult population), to which the Management Committee (Végrehajtóbizottság) is responsible.

Public services. The city has a tradition of good public services, at least in the centre (gas was introduced there for private consumers in 1856 and electricity in 1893), as evidenced by the improvements in public transport and the renovation of the telephone system undertaken in the 1970s. Although the police can be rude and act arbitrarily, they are not overworked: crime, prostitution, and the use of drugs (other than alcohol) are not acute problems. Hospital care and other social services are adequate—particularly in the city centre—although the payment of bribes and gratuities for these services is widespread. Suicide figures are alarmingly high in Budapest, where the most pressing social problem has continued to be the shortage of decent housing.

CULTURAL LIFE

Cultural dominance

The economic hypertrophy of Budapest is compounded by its dominating position in Hungarian culture. Radio and television broadcasting and the film industry are the preserve of the capital, and publishing and the press are nearly so. Writers and poets, traditionally said to be excessively preoccupied with rural life and the peasantry, have nevertheless been drawn to Budapest, as have Hungary's composers: Ferenc Erkel, Ernő Dohnányi, Béla Bartók, and Zoltán Kodály all lived in the capital. The leading grammar schools are concentrated in Budapest, and its universities and colleges attract most of the country's best students. Furthermore, the Hungarian Academy of Sciences and the great majority of its research institutes are in Budapest.

The city also claims the best libraries, museums, art galleries, orchestras, sports facilities, and theatres in Hungary. The music academy, established in 1875 by the pianist and composer Franz Liszt, has acquired international fame. The Opera House was restored to its 19th-century splendour in 1984. The National Theatre, on the other hand, has been awaiting permanent accommodation since it left its building in 1964.

Budapest has more than its share of bookshops; an unparalleled number of hairdressing salons and swimming pools (sometimes the two are combined); and a large variety of thermal baths, including Császár, Lukács, Rác, Király, Rudas, Gellért (all on the Buda side) and Széchenyi and Dagály (in Pest). There are numerous underground hot springs that contain radium and other minerals, and since Roman times bathers have sought them out for their supposed healing properties.

History

EARLY SETTLEMENT AND THE EMERGENCE OF MEDIEVAL BUDA

The Roman settlement

Budapest's location is a prime site for habitation because of its geography, and there is ample evidence of human settlement on the Danube's western side from Neolithic times onward. Two miles north of Castle Hill, in what became Óbuda, a settlement named Ak-Ink ("Ample Water") was established by the Celtic Eravisci. This became Aquincum when the Romans established a military camp and civilian town there at the end of the 1st century AD. Becoming the seat of the province Pannonia Inferior (*c.* AD 106) and then acquiring the status of a municipium (124) and finally a full colony (194), Aquincum grew into a thriving urban centre with two amphitheatres. After the collapse of Roman authority in Pannonia in the early 5th century, some of the large buildings were inhabited by Huns and later by Visigoths and Avars, each group controlling the region for a while.

Kurszán, the Magyar tribal chieftain, probably took up residence in the palace of the former Roman governor at the end of the 9th century. The settlement shifted south to Castle Hill some time after Stephen I of Hungary had established a Christian kingdom in the early 11th century. Buda, for whom the settlement was named, was probably the first constable of the new fortress built on Castle Hill, and the old site to the north became known as Óbuda ("Old Buda"). On the opposite side of the river a Slavonic settlement, Pest (meaning "Lime Kiln," which is also suggested by Ofen, the German name for Buda), was already in existence.

Medieval Buda prospered and declined along with its patron, the Hungarian royal court. The municipality was established by royal charter in 1244, by Béla IV. He bestowed on the citizens of Pest, whose town had been devastated by the Mongols in 1241, the right to settle in full possession of their privileges in the fortified castle. The town administration, based on German law, had been dominated by German burghers before it became reorganized in 1439. At that time parity status was conferred on the Hungarians in municipal government. Buda's preeminence, developed under royal protection, was underlined by its judicial authority (as a higher court) over other free royal towns, although the proximity of the king's court undermined its own self-government. The palace was rebuilt by Matthias I, whose death in 1490 marked the decline of both royal power and the town. The Turks held Buda between 1541 and 1686. After a devastating siege it was liberated by a Christian army organized by the Holy Roman emperor Leopold I. Little of Matthias' Buda survived into the 18th century.

BUDA, ÓBUDA, AND PEST

Both Buda and Pest were recognized by Leopold I as royal free towns in 1703, while Óbuda, a village, belonged to Pest *megye,* the autonomous county that was in the hands of the local Hungarian nobility. In 1720 Buda and Óbuda had a combined population of about 9,600, while that of Pest was only 2,600; but by 1799 Buda had some 24,300 inhabitants to Pest's 29,870, demonstrating that the balance in the size of the two townships had shifted.

Pest, a German commercial centre in Hungary and by then part of the Habsburg empire of Austria, had begun to grow in the late 18th century. Buda, where in the early 18th century only German Roman Catholics were allowed to settle, remained an imperial garrison town and developed once more under the eye of the monarch. A new royal palace was built in the 1760s during the reign of Maria Theresa. The university was moved from Nagyszombat (modern Trnava, Czech.) to Buda in 1777; since 1949 it has been called Loránd Eötvös University. In 1783 Joseph II turned Buda into the country's administrative centre; that same year the Curia (High Court) was moved to Buda, and the university was transferred to Pest. For centuries floods were a serious problem, and one in 1838 took a particularly heavy toll: more than half of the houses in Pest were destroyed, and Buda suffered as well.

The character of Buda under the Habsburgs remained aristocratic and distinctly alien. Pest, into which the gentry and intelligentsia moved, became wedded to the national cause; the Széchenyi Chain Bridge, linking Buda with Pest, was a metaphor for unity. The town of Pest was still partly German, but the nobility of Pest *megye* led the campaign for Hungarian home rule. After the outbreak of revolution in Pest in March 1848, a Hungarian ministry, transferred from Pozsony (modern Bratislava, Czech.) and responsible to the Diet, was established there. In the ensuing civil war Buda was besieged in May 1849 by the revolutionary army of the patriot Lajos Kossuth. Repression followed the revolution until 1867, when the country, which became Austria-Hungary the following year, was placed under the Dual Monarchy. Governments were established in Vienna and Pest.

THE MODERN CITY

Amalgamation

"Buda-Pest" was created in 1872 when Pest, Buda, and Óbuda were united into a single municipal borough comprising 10 districts. Within a year, however, the hyphen had disappeared from the name of the new capital, even in official documents. Between 1873, when the law promulgating the unification went into effect, and the communist takeover after World War II, Budapest enjoyed self-government.

Its governing body, the City Council, consisted of 400 members elected by the districts. The influence of wealth was ensured by a provision of the law (Prussian in origin) that half of the council was to be elected from among the 1,200 highest taxpayers of the capital (the so-called *virilists*); while the other half of the council's membership was elected from the rest of the electorate, based on a rather narrow franchise. Property owners thus played an important role in the government of the city: aristocrats, grain and wine merchants, German burghers, a few solicitors, builders and architects, and later bankers and industrialists all went into real estate and became *virilists.*

Budapest dominated national politics. The Belváros was the constituency of Ferenc Deák, the creator of the Dual Monarchy system. After his death in 1876 the constituency was inherited by leading politicians.

After unification the spectacular growth of the city began in earnest. Baroque and Neoclassical Pest was ruthlessly sacrificed to the building fever that gripped the city fathers. The two most prominent buildings of the period, the new Parliament in Pest and the rebuilt Buda Castle, standing face-to-face over the Danube, were a powerful reminder of the age-old political conflict between the alien monarchy and the national Parliament. Meanwhile, the character of Budapest underwent a major change because of rapid industrialization. Protective tariffs could not be introduced under the dualist system, and the idea of state support for industry gained ground only in the late 1880s, but the result was runaway growth and the creation of an industrial city in which the work force lived in appalling conditions.

Explosive urban and industrial growth created social tension and the emergence of a working-class movement before World War I. After the disintegration of Austria-Hungary in the autumn of 1918, the National Council, a revolutionary body headed by Count Mihály Károlyi and supported by antiwar radicals and Socialists, took power in Budapest. The following March the Károlyi regime collapsed; communists seized power in the capital and held it for four months while also controlling the central regions of the country. Romanian troops occupied and sacked Budapest before the old social order was restored by a counterrevolutionary army marching into the *bűnös város* ("sinful town") in November 1919.

After the signing of the Treaty of Trianon in 1920, Budapest—now the capital of an independent Hungary—became disproportionately large compared to the rest of the country. Social conditions deteriorated in the interwar years. Buda Castle, where Miklós Horthy, the country's regent, resided after 1920, played a fateful role in World War II. Hungary, a satellite of Germany in the war, was occupied by the German army in March 1944. In October Horthy sued for a separate peace and renounced the German alliance, but the coup failed. Between the end of December and the middle of February 1945, German army units put up a fierce resistance to the Soviet advance. In the process, the fighting reduced Castle Hill to a large, sprawling ruin; more than a quarter of the city's buildings and factories were destroyed or damaged, and all the bridges were blown up. It took years for the town to recover from this devastation, but, as the centre of a planned economic system, Budapest resumed its rapid growth in the postwar years.

World War II

The city was enlarged considerably in 1950 when seven satellite towns and 16 villages on its outskirts were merged in its territory. These were then divided into the present 22 administrative districts. One of the most momentous events in the city's postwar history was the revolution in October 1956. Once again the country followed the capital. For the first time a communist government was overthrown by the people. The "seven days that shook the Kremlin" ended in tragedy: the Soviet army retook Budapest, the town once more suffering damage, and the country was forced back into the communist mold. After bloody reprisals the city settled down to a role of providing a model for "goulash" (*i.e.,* market) socialism and for being a shop window to the West. Budapest again became the focus of national political drama in the late 1980s, when Hungary led the reform movement in eastern Europe that broke the communist monopoly on political power and ushered in the possibility of multiparty politics.

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The Buddha and Buddhism

Buddhism is a Pan-Asian religion and philosophy that has played a central role in the spiritual, cultural, and social life of the Eastern world and during the 20th century has spread to the West. This article surveys Buddhism from its origins in the teachings of the Buddha Gautama to its elaboration in various schools, sects, and regional developments.

Ancient Buddhist scripture and doctrine developed primarily in two closely related literary languages of ancient India, Pāli and Sanskrit. In this article, Pāli and Sanskrit words that have gained some currency in English are treated as English words and are rendered in the form in which they appear in English-language dictionaries. Exceptions occur in special circumstances—as, for example, in the case of the Sanskrit term *dharma* (Pāli: *dhamma*), which has meanings that are not usually associated with the English "dharma." Pāli forms are given in the sections that deal with Buddhists whose primary sacred language was Pāli (including discussions of the life and teaching of the Buddha, which are reconstructed on the basis of Pāli texts). Sanskrit forms are given in the sections that deal with Buddhists whose primary focus was on Sanskritic traditions.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 823, and the *Index.*

The article is divided into the following sections:

The foundations of Buddhism

THE CULTURAL CONTEXT

Buddhism came into being in northeastern India during the period from the late 6th century to the early 4th century BC, a period of great social change and intense religious activity. There is disagreement among scholars about the dates of the Buddha's birth and death. Most scholars in Europe, the United States, and India believe that the historical Buddha lived from about 563 to about 483 BC. Many others, especially in Japan, believe that he lived about 100 years later (from about 448 to 368 BC).

At this time in India, many were no longer content with the external formalities of Brahmanic (Hindu high-caste) sacrifice and ritual. In northwestern India there were ascetics who tried to go beyond the Vedas (Hindu sacred scriptures). In the literature that grew out of this movement, the Upanishads, a new emphasis on renunciation and transcendental knowledge can be found. But northeastern India, which was less influenced by the Aryans who had developed the main tenets and practices of the Vedic Hindu faith, became the breeding ground of many heterodox sects. Society in this area was troubled by the breakdown of tribal unity and the expansion of several petty kingdoms. Religiously, this was a time of doubt, turmoil, and experimentation.

A proto-Sāṃkhya sect (a Hindu school founded by Kapila) was already well-established in the area. New sects abounded, including various kinds of skeptics (*e.g.,* Sañjaya Belaṭṭhiputta), atomists (*e.g.,* Pakudha Kaccāyana), materialists (*e.g.,* Ajita Kesakambalī), and antinomians (*i.e.,* those against rules or laws; *e.g.,* Pūraṇa Kassapa). Among the most important sects to arise at the time of the Buddha were the Ājīvikas (Ājīvakas), who emphasized the rule of fate (*niyati*), and the Jainas, an ascetic movement stressing the need to free the soul from matter. Though the Jainas, like the Buddhists, have often been regarded as atheists, their beliefs are actually more complicated. Unlike early Buddhists, both the Ājīvikas and Jainas believed in the permanence of the elements that constitute the universe, as well as the existence of the soul.

Despite the bewildering variety of religious communities, many shared the same vocabulary—nirvana (transcendent freedom), atman ("self," or "soul"), yoga ("union"), karma ("causality"), Tathāgata ("Thus-Gone," or "He Who Has Thus Attained"), buddha ("enlightened one"), samsara ("eternal recurrence," "becoming"), and *dhamma* ("rule," or "law")—and most were based on the practice of yoga. According to tradition, the Buddha himself was a yogi—that is, a miracle-working ascetic.

Buddhism, like many of the sects that developed in northeastern India at the time, was constituted by the presence of a charismatic teacher, by the teachings this leader promulgated, and by a community of adherents that was often made up of renunciant members and lay supporters. In the case of Buddhism this pattern became the basis for the Triratna—the "Three Jewels" of Buddha (the teacher), *dharma* (the teaching), and sangha (the community)—in which Buddhists have traditionally taken refuge.

In the centuries following the founder's death, Buddhism developed in two directions. One, usually called Theravāda by its present-day adherents, remained relatively faithful to what it considered to be the true tradition of the Buddha's teachings. The other is called Mahāyāna, "the means of salvation available to a larger number of people," by its followers, who call the first Hīnayāna, "the means of salvation restricted to a smaller number of people" (or simply the greater and lesser vehicles).

Theravāda and Mahāyāna

In its spread, Buddhism influenced the currents of thought and religion in other countries. In response to the diverse religious aspirations of the various Buddhist communities, the strict law of karma was modified to accommodate new emphases on the efficacy of ritual actions and various forms of devotional practice. Finally there developed in India a movement called Vajrayāna, or Esoteric Buddhism, the aim of which was to obtain liberation more speedily. This movement was influenced by gnostic and magical currents pervasive at that time.

For all the discussion on the two paths of salvation—the gradual and the instant—and the various ways of interpreting the key Mahāyāna concepts of the "void" and the mind-element, the ethics remain fundamentally the same. The monastic organizations suffered the influence of diverse historical situations, but the basic structure remains intact. The Buddha, the original teacher, is always recognized as the revealer of Buddhist truth. In the later doctrines, his preaching is not just that given to his first disciples: he multiplies himself in numberless epiphanies—all manifestations of a single immutable reality—and he emphasizes the certainty of the void and the relativity of all appearances.

In spite of these vicissitudes, Buddhism did not negate its basic principles. Instead they were reinterpreted, rethought, and reformulated, bringing to life an immense literature. This literature includes the Pāli *Tipiṭaka* ("Three Baskets"; three collections of the Buddha's teaching) and the commentaries on it; these were preserved by adherents of the Theravāda tradition. It also includes many sutras and tantras that have been recognized by the followers of the Mahāyāna and Tantric Buddhist traditions as *Buddhavacana,* "the word of the Buddha," along with commentaries on these texts. Consequently, from the first sermon of the Buddha at Sārnāth to the most recent derivations, there is an indisputable continuity—a development or metamorphosis around a central nucleus—by virtue of which Buddhism is differentiated from other religions.

(Gi.T./J.M.K./F.E.R.)

THE LIFE OF THE BUDDHA GOTAMA

The term buddha, literally meaning "awakened one" or "enlightened one," is not a proper name but rather a title, such as messiah (the Christ). Thus, the term should be accompanied by an article, such as "the Buddha" or "a buddha" (because of a belief that there will be innumerable buddhas in the future as there have been in the past). The Buddha who belongs to the present world era was born into the Gotama (in Pāli), or Gautama (in Sanskrit), clan and is often referred to as Gotama. When the term the Buddha is used, it is generally assumed that it refers to Gotama the Buddha.

According to virtually all Buddhist traditions, the Buddha lived many lives before his birth as Gotama; these previous lives are described in stories called Jātakas that play an important role in Buddhist art and education. Most Buddhists also affirm that the Buddha's life was continued in his teachings and his relics. The following account, however, focuses on the Buddha's "historical" life from his birth as Gotama to his death some 80 years later.

Sources

The version of the story presented here is based on the Pāli *Tipiṭaka,* which is recognized by scholars as the earliest extant record of the Buddha's discourses, and on the later Pāli commentaries. The style and technique of these ancient texts, followed in this biography, provide a record—sometimes symbolic, sometimes legendary, and always graphic—of the life of the revered Teacher. Just as there has been a vigorous search for the "historical Jesus" by Christian and other Western-oriented scholars, so also among some Western Orientalists there has been a scholarly search for the "historical Buddha," the history of whom the Buddhists themselves never questioned and which had never interested them as a historical problem. This section concentrates on Gotama the Enlightened One as depicted in the Buddhist scriptures and legends that developed about the man, his teachings, and his activities.

Birth and early life. The Buddha was born in the 6th or 5th century BC in the kingdom of the Sakyas, on the borders of present-day Nepal and India. As the son of Suddhodana, the king, and Mahāmāyā, the queen, the Buddha thus came from a Khattiya family (*i.e.,* the warrior caste or ruling class).

The story of the Buddha's life, however, begins with an account of a dream that his mother Mahāmāyā had one night before he was born: a beautiful elephant, white as silver, entered her womb through her side. Brahmans (Vedic priests) were asked to interpret the dream, and they foretold the birth of a son who would become either a universal monarch or a buddha. Ten lunar months after the conception, the queen and her retinue left Kapilavatthu, the capital of the Sakya kingdom, on a visit to her parents in Devadaha. She passed through Lumbinī, a park that was owned jointly by the people of both cities. There, she gave birth to the Buddha in a curtained enclosure in the park on the full-moon day of the month of Vesākha (May). The purported site of his birth, now called Rummindei, lies within the territory of Nepal. A pillar placed there in commemoration of the event by Aśoka, a 3rd-century-BC Buddhist emperor of India, still stands.

Immediately upon hearing of the birth of the Buddha, the sage Asita (also called Kāḷa Devala), who was King Suddhodana's teacher and religious adviser, went to see the child. From the auspicious signs on the child's body, Asita recognized that this child would one day become a buddha, and he was overjoyed and smiled. Because he was very old, however, he grew sad and wept, knowing that he would not remain alive to see the child's subsequent Enlightenment. Suddhodana, because of this strange display of alternate emotions, was concerned about possible dangers to the child, but Asita explained why he had first smiled and then wept and reassured the king about the child's future. Both the sage and the king then worshiped the child.

On the fifth day after birth, for the name-giving ceremony, 108 Brahmans were invited, among whom eight were specialists in interpreting bodily marks. Of these eight specialists, seven predicted two possibilities: if the child remained at home, he would become a universal monarch; if he left home, he would become a buddha. But Koṇḍañña, the youngest of the eight, predicted that he would definitely become a buddha. Later, this same Koṇḍañña became one of the Buddha's companions and was one of his first five disciples. The child was given the name Siddhattha (Sanskrit: Siddhārtha), which means "one whose aim is accomplished."

On the seventh day after his birth, his mother died, and the child was brought up by her sister Mahāpajāpatī Gotamī, Suddhodana's second consort.

Signs of the Buddha's eventual career

A significant incident in the Buddha's boyhood is recorded in ancient Pāli commentaries. One day, the little Siddhattha was taken to the state plowing festival, in which the king, with his ministers and the ordinary farmers, took part, according to the custom of the Sakyas. The boy was left with his nurses in a tent under a *jambu* tree. The nurses, attracted by the festivities, left the prince alone in the tent and went out to enjoy themselves. When they returned, they found the boy seated cross-legged, absorbed in a trance (Pāli: *jhāna*; Sanskrit: *dhyāna*). The king was immediately informed and saw his little son in the posture of a yogi (a practitioner of psychological, physiological meditation techniques). Upon seeing his son sitting in this fashion, he worshiped the child a second time. Many years later the Buddha himself, in one of his discourses (the *Mahā-Saccaka-sutta,* "The Great Discourse to Saccaka," of the Pāli *Majjhima Nikāya,* or the "Collection of the Middle Length Sayings of the Buddha"), briefly mentions his attaining to the first *jhāna* under the *jambu* tree.

The young prince was brought up in great luxury, and his father, always worried that his son might leave home to become a wandering ascetic as the Brahmans had predicted, took every care to influence him in favour of a worldly life. According to the *Aṅguttara Nikāya* ("Collection of the Gradual Sayings of the Buddha"), the Buddha himself is reported to have said later about his upbringing:

> Bhikkhus [monks], I was delicately nurtured, exceedingly delicately nurtured, delicately nurtured beyond measure. In my father's residence lotus-ponds were made: one of blue lotuses, one of red and another of white lotuses, just for my sake.... Of Kāsi cloth was my turban made; of Kāsi my jacket, my tunic, and my cloak.... I had three palaces: one for winter, one for summer and one for the rainy season.... in the rainy season palace, during the four months of the rains, entertained only by female musicians, I did not come down from the palace.

At the age of 16, Siddhattha married his cousin, a princess named Yasodharā, also 16 years old. Although Suddhodana tried his utmost to make Siddhattha content by providing him with luxury and comfort, the young prince's thoughts were generally elsewhere, occupied with other concerns.

The Four Signs. The turning point in the prince Siddhattha's life came when he was 29 years old. One day, while out driving with his charioteer, he saw "an aged man as bent as a roof gable, decrepit, leaning on a staff, tottering as he walked, afflicted and long past his prime." The charioteer, questioned by the prince as to what had happened to the man, explained that he was old and that all men were subject to old age. The prince, greatly perturbed by this sight, went back to the palace and became absorbed in thought. Another day, again driving with his charioteer, he saw "a sick man, suffering and very ill, fallen and weltering in his own excreta." Because Siddhattha was perturbed, the charioteer explained, as before, that this was a sick man and that all men are subject to sickness. On a third occasion the prince saw a dead body and again the charioteer provided the explanation. Finally, Siddhattha saw "a shaven-headed man, a wanderer who has gone forth, wearing the yellow robe." Impressed with the man's peaceful demeanour, the prince decided to leave home and go out into the world to discover the reason for such a display of serenity in the midst of misery.

On his way back to the palace after seeing the yellow-robed ascetic, Siddhattha received the news of the birth of his son, whom he named Rāhula, meaning "Fetter" or "Bond."

The Great Renunciation. Upon receiving this news, the prince decided to make what is known as the Great Renunciation: to give up the princely life and become a wandering ascetic. Waking up in the middle of the night, he ordered Channa, his charioteer and companion, to saddle his favourite horse, Kanthaka, and went to the bedchamber to have a last look at his sleeping wife and their son. He did not enter the chamber for fear of awakening his wife, which would be a sure obstacle to his plan. He thought he would one day come to see them again.

Beginning of the quest for the truth

That night Siddhattha left the city of Kapilavatthu, accompanied by Channa. By dawn he had crossed the Anomā River. He then gave all his ornaments to Channa, assumed the guise of an ascetic, and sent Channa and Kanthaka back to his father.

As an ascetic, Gotama went south, where centres of learning and spiritual discipline flourished, and arrived at Rājagaha (modern Rājgīr), the capital of the Magadha kingdom. Bimbisāra, the king of Magadha, was impressed by the handsome appearance and the serene personality of this strange ascetic and visited him when he was seated at the foot of a hill. The king, after he discovered that the ascetic was a former prince, offered him every comfort and suggested that he should stay with him to share his kingdom. Gotama, however, rejected the king's offer, saying that he had no need of those things that he had renounced and that he was in search of truth. Bimbisāra then requested that, when Gotama obtained the Enlightenment, he return to visit Rājagaha again, to which Gotama agreed.

The search for the truth. Leaving Rājagaha, Gotama went in search of teachers to instruct him in the way of truth. Two of them the Buddha himself mentioned by name in several discourses. He first went to Āḷāra Kālāma, a renowned sage, and expressed his wish to follow Āḷāra's system; Āḷāra gladly accepted Gotama as his pupil. Gotama studied and rapidly mastered Āḷāra's whole system and then asked his teacher how far the master himself had realized that teaching. Āḷāra told him that he had attained the "sphere of no-thing." Gotama soon attained the same mystical state himself. Āḷāra admitted that that state was the highest he could teach and declared that Gotama and himself were now equals in every respect—in knowledge, practice, and attainment—and invited the Sakyan ascetic to guide, along with him, the community of his disciples. The Buddha later spoke of this occasion in a *sutta:* "In this way did Āḷāra Kālāma, my teacher, set me, his pupil, on the same level as himself and honoured me with the highest honour." Gotama, however, was not satisfied with attaining the sphere of no-thing, though it was a very high mystical state. He was in quest of absolute truth, nirvana, and thus he left Āḷāra Kālāma.

He then went to Uddaka Rāmaputta, another great teacher, who taught him to attain the "sphere of neither-perception-nor-nonperception," a higher mystical state than the sphere of no-thing. Gotama, however, was not satisfied with this either, and he continued his search for the truth.

Traveling through the Magadha country, Gotama arrived at a village called Senānigama, near Uruvelā, and, according to his own words, found "a beautiful stretch of land, a lovely woodland grove, and a clear flowing river with a pleasant ford, and a village for support close by." He was joined there by a group of five ascetics, among whom was Koṇḍañña, the Brahman who had predicted at the name-giving ceremony that the child Siddhattha would definitely become a buddha one day.

The Buddha's practice of asceticism

Gotama's real struggle in his search for the truth began in the area around Uruvelā, near modern Gayā. Here, for nearly six years, he practiced various severe austerities and extreme self-mortifications. These austerities were vividly described in several discourses attributed to the Buddha himself (*e.g.,* in the *Majjhima Nikāya*). What he looked like and what happened to him is described in the following words from the ancient text:

> Because of so little nourishment, all my limbs became like some withered creepers with knotted joints; my buttocks like a buffalo's hoof; my back-bone protruding like a string of balls; my ribs like rafters of a dilapidated shed; the pupils of my eyes appeared sunk deep in their sockets as water appears shining at the bottom of a deep well; my scalp became shriveled and shrunk as a bitter gourd cut unripe becomes shriveled and shrunk by sun and wind; . . . the skin of my belly came to be cleaving to my back-bone; when I wanted to obey the calls of nature, I fell down on my face then and there; when I stroked my limbs with my hand, hairs rotted at the roots fell away from my body.

Many later representations of the Buddha portray him in this emaciated state.

As a consequence of these severe bodily austerities, Gotama became so weak that he once fainted and was believed by some to be dead. From these experiences, he realized that such mortifications could not lead him to what he sought; he therefore changed his way of life and again began to eat proper amounts of food.

His five companions, who had much faith in him, were disappointed at his rejection of extreme asceticism and left him in disgust. Gotama thus remained alone in Uruvelā, regained his health and strength, and then followed his own path to Enlightenment.

The Great Enlightenment. One morning, seated under a banyan tree, Gotama accepted an offering of a bowl of milk rice from Sujātā, the daughter of the landowner of the village of Senānigama. This was his last meal before his Enlightenment. He spent the day in a grove of sal trees and in the evening went to the base of a pipal tree (*Ficus religiosa*), now known as the bodhi, or bo, tree, and sat cross-legged, determined not to rise without attaining Enlightenment.

The struggle with Māra

At that point, the greatest of Gotama's struggles began: Māra, the evil one, the tempter who is the lord of the world of passion, determined to defeat him and prevent him from attaining Enlightenment; he approached Gotama with his hideous demonic hordes. Gotama, however, sat unmoved in meditation, supported only by the 10 paramitas ("great virtues") that he had perfected during innumerable past lives as a bodhisattva ("buddha-to-be") in order to attain Enlightenment. (In order to attain buddhahood, all bodhisattvas [*i.e.,* those who aspire to become buddhas] have to perfect, during innumerable lives, these 10 paramitas: charity, morality, renunciation, wisdom, effort, patience, truth, determination, universal love, and equanimity.) Māra was thus vanquished and fled headlong with his armies of evil spirits.

The battle with Māra is graphically described in ancient Buddhist texts and depicted in paintings on the walls of Buddhist temples. In the *Padhānasutta* ("Discourse on the Exertion") of the Pāli *Suttanipāta,* one of the earliest texts, the Buddha states that, when he was practicing austerities by the Nerañjarā River in Uruvelā, Māra approached him, speaking such words as: "You are emaciated, pale, you are near death. Live, Sir, life is better. Do meritorious deeds. What is the use of striving?" After some preliminary words, Gotama replied:

> Lust is your first army; the second is dislike for higher life; the third is hunger and thirst; the fourth is craving; the fifth is torpor and sloth; the sixth is fear (cowardice); the seventh is doubt; the eighth is hypocrisy and obduracy; the ninth is gains, praise, honour, false glory; the tenth is exalting self and despising others. Māra, these are your armies. No feeble man can conquer them, yet only by conquering them one wins bliss. I challenge you! Shame on my life if defeated! Better for me to die in battle than to live defeated.

Māra, overcome with grief, disappeared.

The realization of the Four Noble Truths

Having defeated Māra, Gotama spent the rest of the night in deep meditation under the tree. During the first part of the night he gained the knowledge of his former existences. During the second part of the night he attained the "superhuman divine eye," the power to see the passing away and rebirth of beings. In the last part of the night he directed his mind to the knowledge of the destruction of all cankers and defilements and realized the Four Noble Truths. In words attributed to the Buddha himself: "My mind was emancipated, . . . Ignorance was dispelled, science (knowledge) arose; darkness was dispelled, light arose."

Thus Gotama, at the age of 35, attained the Enlightenment, or Awakening, and became a supreme buddha during the night of the full-moon day of the month of Vesākha (May) at a place now called Bodh Gayā (Pāli and Sanskrit: Buddhagayā).

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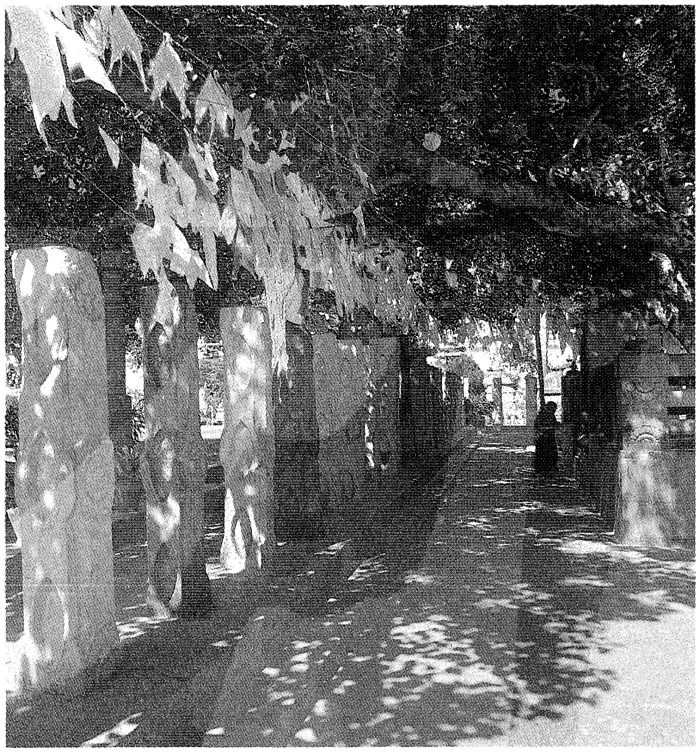


Figure 1: Prayer flags and pilgrim under the bodhi tree at Bodh Gayā, India, the site of the Buddha's Enlightenment.

Contemplation on the truth. After his Enlightenment the Buddha spent several weeks (five or seven weeks according to different accounts) in Uruvelā, meditating on the various aspects of the *dhamma* that he had realized, particularly on the most important and difficult doctrine of causal relations, known as the dependent origination or the conditioned genesis (*paṭicca-samuppāda*). This doctrine views everything as relative and interdependent and teaches that there is no eternal, everlasting, unchanging, permanent, or absolute substance, such as the soul, the self, or the ego, within or without man.

Four weeks after his Enlightenment, seated under a banyan tree, the Buddha is reported to have thought to himself: "I have realized this Truth which is deep, difficult to see, difficult to understand . . . comprehensible by the wise. Men who are overpowered by passion and surrounded by a mass of darkness cannot see this Truth which is against the current, which is lofty, deep, subtle and hard to comprehend."

With these thoughts in mind, the Buddha hesitated to try to explain to the world the truth that he had just realized. At this point, according to the tradition, the Brahman Sahampati intervened in order to convince the Buddha to accept his vocation as a teacher. This great Brahmanic deity set forth for him an image of a lotus pond: in a lotus pond there are some lotuses still under water; there are others that have risen only up to the water level; and there are still others that stand above water and are untouched by it. In a similar way, in this world there are people of different levels of development. Thus challenged, the Buddha determined to proclaim the insight he had gained.

At the outset he faced the problem of choosing those who would be the first to hear him preach the *dhamma.* He first thought of his two former teachers, Āḷāra Kālāma and Uddaka Rāmaputta, but they had died by this time. He then thought of the five companions who had left him and were now staying in Isipatana near Bārāṇasī (Benares; now Vārānasi) and decided to go there.

On meeting the five ascetics, the Buddha told them that now he was an arhat, a "perfected one" (Pāli: *arahant*), a "fully awakened one" (*sammāsambuddha*), that he had realized the "immortal" (*amata*), and that he wished to instruct and teach them the *dhamma.* They replied to him:

> But, Reverend Gotama, even by all that conduct, that practice, that austerity, you did not realize this supreme knowledge, this supreme state. So how can you now realize it when you live in abundance, when you have given up striving and have reverted to a life of abundance?

The Buddha denied that he had given up striving and that he had reverted to a life of abundance. He requested again that they listen to him. Again, however, they replied in a similar manner. A third time the Buddha repeated what he had said and asked them to listen to him, and they repeated their remark.

The Buddha then asked them a question: "Do you admit that I have never spoken anything like this before?" They were struck by such straightforwardness and knew how sincere and earnest he was. Convinced that he had attained what he claimed to have attained, they no longer addressed the Buddha as "Reverend Gotama" but changed their attitude toward him and answered him: "Lord, you have not." The Buddha then delivered to them his first sermon, known as the *Dhammacakkappavattana-sutta* ("Sermon on Setting in Motion the Wheel of Truth"), at Isipatana, now called Sārnāth. An ancient stupa (a building containing a religious relic) still marks the spot where this event supposedly occurred.

The teaching of moderation and the Eightfold Path

The substance of this *sutta* is as follows: a man who has left home and gone forth should not follow two extremes, namely self-indulgence and self-mortification. Avoiding these two extremes, the Tathāgata ("He Who Has Thus Attained"—*i.e.,* the Buddha) has discovered the middle path leading to vision, to knowledge, to calmness, to awakening, to nirvana. This middle path is known as the Noble Eightfold Path consisting of right view, right thought, right speech, right action, right mode of living, right endeavour, right mindfulness, and right concentration. The First Noble Truth is that man's existence is *dukkha,* full of conflict, dissatisfaction, sorrow, and suffering. The Second Noble Truth is that all this is caused by man's selfish desire—*i.e.,* craving or *taṇhā,* "thirst." The Third Noble Truth is that there is emancipation, liberation, and freedom for human beings from all this, which is nirvana. The Fourth Noble Truth, the Noble Eightfold Path, is the way to this liberation.

The founding of the sangha. At the end of the sermon, these five ascetics, the Buddha's first disciples, were admitted by him as bhikkhus (monks) and became the first members of the sangha ("community," or "order"). A few days later, this sermon was followed by the *Anattalakkhaṇa-sutta,* dealing with the doctrine of no-self, at the conclusion of which all five bhikkhus became arhats ("perfected ones").

The Buddha spent about three months in the Vārānasi/Benares region. During this period an important and influential wealthy young man named Yasa became his disciple and entered the order. His father and mother, along with his former wife, also were converted. They were the first lay disciples to take refuge in the "Triple Jewel": the Buddha, the *dhamma,* and the sangha. Later, four of Yasa's close friends followed his example and entered the order. Enthusiasm for this new movement became so impelling that 50 of their friends also joined them in the sangha. All these became arhats in due course, and the Buddha soon had 60 disciples who were perfected ones.

The Buddha addressed this group in the following words and sent them out into the world to spread his message of peace, compassion, and wisdom:

> Bhikkhus, I am freed from all fetters, both divine and human. You, too, are freed from all fetters, both divine and human. Wander forth, bhikkhus, for the good of the many, for the happiness of the many, out of compassion for the world.... Let not two of you go by one road [*i.e.,* go in different directions]. Teach the *Dhamma* which is good at the beginning, good in the middle, and good at the end.... There are people who will understand the *Dhamma.* I, too, will go to Uruvelā to teach the *Dhamma.*

The 60 disciples went in various directions to spread the teaching of the Buddha. The Buddha himself set out for Uruvelā. On the way he converted 30 young men, who then entered the order. In the region of Uruvelā he also converted three leading ascetics along with a large number of their disciples. To these ascetics, formerly known as "those with matted hair" (*jaṭilas*), the Buddha delivered the famous "Fire Sermon" (the *Adittapariyāya-sutta*), which states that all man's existence is burning with the fire of lust, the fire of hate, and the fire of delusion.

From Uruvelā the Buddha went on to Rājagaha, the capital of Magadha, fulfilling his promise to visit King Bimbisāra after his Enlightenment. Many people, including the king, became his lay disciples. The king offered his park, Veḷuvana, as a monastery site to the Buddha and his order. During this visit a very important event that had far-reaching effects took place: Sāriputta and Moggallāna, two Brahmanic ascetics who later became the Buddha's two chief disciples, joined the order. Sāriputta had first heard of the Buddha and his new teaching from Assaji, one of the original 60 disciples. At the request of his father, the Buddha visited Kapilavatthu with a large number of his disciples. In that city, where as prince he had lived in great splendour and luxury, he went about begging for his food from house to house. His father, King Suddhodana, was grieved and upset by this, but, upon learning that this was the custom of all buddhas, he conducted the Blessed One and his disciples to eat a meal at the palace.

Conversion of members of the Buddha's family

All the ladies of the court went to him to offer reverence, except his former wife, Yasodharā. She refused, saying that the Blessed One himself would come to her if he thought she had any virtue in her and that she would then worship him. The Buddha, with his two chief disciples and the king, went to see her in her apartment. She fell at his feet, clasped his ankles with her hands, and put her head on his feet.

The Buddha's father, his aunt Mahāpajāpatī, Yasodharā, and large numbers of Sakyans (who were fellow members of the Gotama clan) became his followers. On the following day he ordained his half-brother Nanda and a few days later his son, Rāhula. All this troubled the old king so much that he asked the Buddha to lay down a rule that no son should be ordained without the consent of his parents. Accordingly, the rule was formulated, and it continues to be followed by the sangha.

Anāthapiṇḍika, a banker of Sāvatthi (modern Śrāvasti), the capital of Kosala kingdom, had met the Buddha at Rājagaha and had become deeply devoted to him. He invited the Blessed One to his city, where he built for him the famous monastery at Jetavana. This monastery in Sāvatthi became the virtual headquarters of the Buddha's activities. There he spent most of his time and delivered most of his sermons. The Buddha and his new teaching became so popular that monasteries were built for him and his sangha in almost all the important cities in the valley of the Ganges, and the number of his followers among all classes of people increased rapidly.

The order of nuns, *bhikkhunī-saṅgha,* was instituted after some hesitation. Ānanda, the Buddha's cousin and later his chief attendant and constant companion, pleaded with the Master on behalf of women. The Buddha's own aunt Mahāpajāpatī Gotamī and her friends were the first women to enter the order.

Members of some hostile sects, who became jealous of the Buddha's success and popularity, made several attempts to vilify him.

Devadatta, one of the Buddha's cousins, an ambitious man of ability and guile, was his rival from early days. He too joined the order but was never sincerely devoted to the Master. He became popular and influential with some people, however, and, about eight years before the Buddha's death, Devadatta conceived the idea of becoming the Buddha's successor and suggested to him that the leadership of the sangha should be handed over to him in view of the Master's approaching old age. The suggestion, however, was rejected. The Buddha stated that he would not pass on the leadership of the order to anyone, not even to Sāriputta or Moggallāna. Rather, the sangha was to be run in accordance with democratic principles. Its constitution was to be the vinaya ("discipline"), rules that the Buddha himself had laid down to guide the spiritual and material life of the individual monks and nuns and to regulate the structure and dynamics of monastic life.

After being rebuffed in this way, Devadatta vowed vengeance. He made three cleverly designed attempts on the life of the Buddha, all of which failed. Devadatta next tried to bring about a schism in the sangha, taking with him a group of newly ordained monks to establish a separate community. All those who were misled by Devadatta, however, were later persuaded to go back to the Master by Sāriputta and Moggallāna. After this event Devadatta became seriously ill and died after about nine months of illness.

The death of the Buddha. After the Buddha had trained learned, well-disciplined followers and his mission was fulfilled, at the age of 80, with a group of monks, he set out on his last journey, from Rājagaha toward the north. As usual, he passed in leisurely fashion through cities, towns, and villages, teaching the people on his way and stopping wherever he wished.

The Buddha's attempt to recover his health

In due course he arrived at Vesālī, the capital city of the Licchavis. The Buddha spent that rainy season not in the park in Vesālī, which had just been donated to him by Ambapālī, the celebrated courtesan of that city, but in an adjoining village called Beluva. There the Buddha became seriously ill. He thought, however, that it was not right for him to die without preparing his disciples, who were dear to him. Thus, with courage, determination, and will, he bore all his pains, got the better of his illness, and recovered; but his health was still poor.

After the Buddha's recovery, Ānanda, his most devoted attendant, went to his beloved Master and said:

> Lord, I have looked after the health of the Blessed One. I have looked after him in his illness. But at the sight of his illness, the horizon became dim to me, and my faculties were no longer clear. Yet there was one little consolation: I thought the Blessed One would not pass away until he had left instructions concerning the Order of the Sangha.

The Buddha, full of compassion and feeling, replied:

> Ānanda, what does the Order of the Sangha expect from me? I have taught the *dhamma* without making any distinction as to exoteric and esoteric. With regard to the Truth, the Tathāgata has nothing like the 'closed fist of a teacher' (*ācariya-muṭṭhi*), who keeps something back. Surely, Ānanda, if there is anyone who thinks that he will lead the sangha and that the sangha should depend on him, let him set down his instructions. But the Tathāgata has no such idea. Why should he then leave instructions concerning the sangha? I am old now, Ānanda . . . eighty years old. As a worn-out cart has to be kept going by repairs, so, it seems to me, the body of the Tathāgata can only be kept going by repairs.... Therefore, Ānanda, dwell by making yourselves your island, making yourselves, not anyone else, your refuge; making the *dhamma* your island, the *dhamma* your refuge, nothing else your refuge.

Later the Buddha told Ānanda that he had decided to die after three months and asked him to assemble in the hall at Mahāvana all the monks who were at that time residing in the neighbourhood of Vesālī. At this meeting, the Buddha advised the monks to follow what he had taught them and to spread it abroad for the good of the many, out of compassion for the world. He then announced that he had decided to die after three months.

Leaving Vesālī, the Buddha gazed at the city in which he had stayed on many occasions and said: "This will be the last time, Ānanda, that the Tathāgata will behold Vesālī. Come, Ānanda, let us proceed."

Stopping at several villages and townships, the Buddha eventually arrived at Pāvā and stayed in the park of Cunda the goldsmith, who was already one of his devoted followers. At his invitation the Buddha and the monks went to his house for a meal. Cunda had prepared, besides various delicacies, a dish called *sūkara-maddava.* This is interpreted in the ancient Pāli commentaries in several ways: (1) as pork (this is generally accepted), (2) as bamboo sprouts trodden by pigs, (3) as a kind of mushroom growing in a spot trodden by pigs, (4) as a rice pudding rich with the essence of milk, or (5) as a special preparation (an elixir?) intended by Cunda to prolong the Buddha's life. Whatever it might have been, the Buddha asked Cunda to serve him with *sūkara-maddava* and to serve the bhikkhus with other dishes. At the end of the meal, the Buddha requested Cunda to bury in a hole whatever was left of the *sūkara-maddava,* saying that only a Tathāgata would be able to assimilate it. This was the Buddha's last meal.

After it the Buddha became sick and suffered violent pains but bore them without complaint. He set out for Kusinārā, accompanied by Ānanda and other monks. Explaining that he was tired, he stopped and rested in two places. On the way, the Buddha said to Ānanda:

> Now it may happen, Ānanda, that someone should stir up remorse in Cunda by saying that the Tathāgata died after eating his meal. Any such remorse in Cunda should be dispelled. Tell him, Ānanda, that you heard directly from my mouth that there are two offerings of food which are of equal fruit, of equal profit: the offering of food before the Enlightenment and the offering of food before the *Parinibbāna* (the passing away) of a Tathāgata. Tell him that he has done a good deed. In this way Ānanda, you should dispel any possible remorse in Cunda.

The Buddha arrived at Kusinārā (the modern Kasia, known in Sanskrit as Kuśinagara) toward evening, and, on a couch between two sal trees in the park Upavattana of the Mallas, he "laid himself down on his right side, with one leg resting on the other, mindful and self-possessed." This was the full-moon day of the month of Vesākha (May).

Final conversations with Ānanda and other monks

Ānanda asked the Buddha what they should do with his remains. He told Ānanda they should not occupy themselves with honouring the remains of the Tathāgata but should rather be zealous in their own spiritual development. The lay devotees, he said, would busy themselves with the remains.

Ānanda left the immediate area and cried out: "My Master is about to pass away from me—he who is so kind to me." The Buddha inquired where Ānanda was and, on being told that he was weeping, called to him and said: "No, Ānanda, don't weep. Haven't I already told you that separation is inevitable from all near and dear to us? Whatever is born, produced, conditioned, contains within itself the nature of its own dissolution. It cannot be otherwise." Then, the Master spoke to the monks in praise of Ānanda's wonderful qualities and abilities. The Mallas, in whose realm Kusinārā was located, came with their families to pay homage to the Blessed One. A wandering ascetic named Subhadda asked for permission to see the Buddha, but Ānanda refused, saying that the Blessed One was tired and that he should not be troubled. The Buddha, overhearing the conversation, called Ānanda and asked him to allow Subhadda to see him. After an interview with the Buddha, Subhadda joined the order the same night, thus becoming his last direct disciple.

The Buddha then addressed Ānanda:

> It may be, Ānanda, that to some of you the thought may come: 'Here we have the Word of the Master who is gone; our Master we have with us no more.' But, Ānanda, it should not be considered in this light. What I have taught and laid down, Ānanda, as *Dhamma* (Truth, Doctrine) and as *Vinaya* (Discipline), this will be your Master when I am gone. . . . If the sangha wish it, Ānanda, let them, when I am gone, abolish lesser and minor precepts (rules).

The Buddha next addressed the monks and requested them three times to ask him if they had any doubt or question that they wished clarified, but they all remained silent. The Buddha then addressed the monks: "Then, bhikkhus, I address you now: transient are all conditioned things. Try to accomplish your aim with diligence." These were the last words of the Tathāgata. A week later, his body was cremated by the Mallas in Kusinārā.

A dispute over the relics of the Buddha arose between the Mallas and the delegates of rulers of several kingdoms, such as Magadha, Vesālī, and Kapilavatthu. It was settled by a venerable old Brahman named Doṇa on the basis that they should not quarrel over the relics of one who preached peace. With common consent, the relics were then divided into eight portions to the satisfaction of all. Stupas were built over these relics, and feasts were held commemorating the Buddha.

Assessment of the personality and character of the Buddha. According to the Buddhist tradition, the Buddha was a very handsome man. Caṅkī, a highly respected Brahman leader, is reported to have said that "the recluse Gotama is lovely, good to look upon, charming, possessed of the greatest beauty of complexion, of a sublime colour, a perfect stature, noble of presence." Buddhists came to envision (and later represent) him as one endowed with the 32 bodily characteristics of a *mahāpuruṣa* ("great person").

He had a unique reputation as a superb teacher. His conversion and taming of Aṅgulimāla, a murderer and bandit who was a terror even to Pasenadi, the king of Kosala, is put forward as an example of his great powers and abilities. People who went to see and hear him were fascinated and were so quickly converted to his new teaching that his opponents described him as having some "enticing trick." King Pasenadi is reported to have said that those who went with the idea of confounding the Buddha in debate became his disciples at the end. Full of compassion and wisdom, he is recognized as knowing how and what to teach individual people for their own benefit according to the level of their capabilities.

The Buddha, affectionate and devoted to his disciples, was always inquiring after their well-being and progress. When he was staying in a monastery, he paid daily visits to the sick ward. Once, he himself attended a sick monk neglected by others and made the comment that "he who attends on the sick attends on me."

The Buddha as a social reformer

The Buddha refused to recognize the religious significance of the caste system that was a long-established and respected institution in India and recognized the religious potential of men and women of all social ranks. He also recognized the connection between economic welfare and moral development. Trying to suppress crime through punishment, he said, was futile. Poverty, according to the Buddha, was a cause of immorality and crime; therefore, the economic condition of people should be improved.

He appreciated both natural and physical beauty. On several occasions he was moved aesthetically, as he told Ānanda how delightful certain places were to him. At Vesālī he told the monks that, if they had not seen the devas (gods) of Tāvatiṃsa (Heaven), they should look at the handsome Licchavis, beautifully and elegantly dressed in different colours.

King Pasenadi could not understand how the Buddha maintained such order and discipline in the community of monks, when he, a king, with the power to inflict punishment, could not maintain it as well in his court. The Buddha, however, kept order and discipline on the basis of a mutual love, affection, and respect that exists between teacher and pupil.

Many miraculous powers were attributed to the Buddha, and he performed a number of miracles during his ministry. At the same time, however, he did not consider magical powers to be of primary importance. Once, when one of his disciples performed a miracle in public, the

Buddha reproached him and laid down a rule that his disciples should not perform miracles before the laity. In his view, the greatest miracle was to explain the truth and to make people recognize its importance.

Behind his philosophy and strict ethics, the Buddha had a quiet sense of humour. A conceited Brahman, who was in the habit of denigrating others, questioned him as to the qualities of a true Brahman. In a list of such high qualities as freedom from evil and purity of heart, the Buddha gently included "not denigrating others."

The portrait of the Buddha, as can be inferred from the lines of the ancient texts, is thus one of a man of great wisdom and great compassion, one who was moved by the spectacle of human suffering and was determined to teach his fellow human beings how that suffering could be confronted and overcome. (Wa.R./F.E.R.)

THE BUDDHA'S MESSAGE

The teaching attributed to the Buddha was transmitted orally by his disciples, prefaced by the phrase "evaṃ me sutaṃ" ("thus have I heard"); therefore, it is difficult to say whether his discourses were related as they were spoken. They usually allude, however, to the place, time, and community where he preached; and there is concordance between various versions. An attempt was made by Buddhist councils in the first centuries after the Buddha's death to establish his true and original teachings.

Suffering, impermanence, and no-self. It may be said that the Buddha based his entire teaching on the fact of human suffering. Existence is painful. The conditions that make an individual are precisely those that also give rise to suffering. Individuality implies limitation; limitation gives rise to desire; and, inevitably, desire causes suffering, since what is desired is transitory, changing, and perishing. It is the impermanence of the object of craving that causes disappointment and sorrow. By following the "path" taught by the Buddha, the individual can dispel the "ignorance" that perpetuates this suffering. The Buddha's doctrine was not one of despair. Living amid the impermanence of everything and being themselves impermanent, human beings search for the way of deliverance, for that which shines beyond the transitoriness of human existence—in short, for enlightenment.

The constituents of reality

According to the Buddha, reality, whether of external things or the psychophysical totality of human individuals, consists in a succession and concatenation of microseconds called *dhamma*s (these "components" of reality are not to be confused with *dhamma* meaning "law" or "teaching"). The Buddha departed from the main lines of traditional Indian thought in not asserting an essential or ultimate reality in things.

Moreover, contrary to the theories of the Upanishads, the Buddha did not want to assume the existence of the soul as a metaphysical substance, but he admitted the existence of the self as the subject of action in a practical and moral sense. Life is a stream of becoming, a series of manifestations and extinctions. The concept of the individual ego is a popular delusion; the objects with which people identify themselves—fortune, social position, family, body, and even mind—are not their true selves. There is nothing permanent, and, if only the permanent deserved to be called the self, or atman, then nothing is self. There can be no individuality without a putting together of components. This is becoming different, and there can be no way of becoming different without a dissolution, a passing away.

To make clear the concept of no-self (anatman), Buddhists set forth the theory of the five aggregates or constituents (*khandha*s) of human existence: (1) corporeality or physical forms (*rūpa*), (2) feelings or sensations (*vedanā*), (3) ideations (*saññā*), (4) mental formations or dispositions (*sankhāra*), and (5) consciousness (*viññāṇa*). Human existence is only a composite of the five aggregates, none of which is the self or soul. A person is in a process of continuous change, with no fixed underlying entity.

Karma. The belief in rebirth, or samsara, as a potentially endless series of worldly existences in which every being is caught up was already associated with the doctrine of karma (Sanskrit: *karman;* literally "act," or "deed") in pre-Buddhist India, and it was generally accepted by both the Theravāda and the Mahāyāna traditions. According to the doctrine of karma, good conduct brings a pleasant and happy result and creates a tendency toward similar good acts, while bad conduct brings an evil result and creates a tendency toward repeated evil actions. This furnishes the basic context for the moral life of the individual.

Some karmas bear fruit in the same life in which they are committed, others in the immediately succeeding one, and others in future lives that are more remote.

The acceptance by Buddhists of the belief in karma and rebirth while holding to the doctrine of no-self gave rise to a difficult problem: how can rebirth take place without a permanent subject to be reborn? Indian non-Buddhist philosophers attacked this vulnerable point in Buddhist thought, and many modern scholars have also considered it to be an insoluble question. The relation between existences in rebirth has been explained by the analogy of fire, which maintains itself unchanged in appearance and yet is different in every moment—what may be called the continuity of an ever-changing identity.

The Four Noble Truths. Awareness of these fundamental realities led the Buddha to formulate the Four Noble Truths: the truth of misery, the truth that misery originates within us from the craving for pleasure and for being or nonbeing, the truth that this craving can be eliminated, and the truth that this elimination is the result of a methodical way or path that must be followed. Thus, there must be an understanding of the mechanism by which a human being's psychophysical being evolves; otherwise, human beings would remain indefinitely in samsara, in the continual flow of transitory existence.

The law of dependent origination. Hence, the Buddha formulated the law of dependent origination (*paṭicca-samuppāda*), whereby one condition arises out of another, which in turn arises out of prior conditions. Every mode of being presupposes another immediately preceding mode from which the subsequent mode derives, in a chain of causes. According to the classical rendering, the 12 links in the chain are ignorance (*avijjā*), karmic predispositions (*sankhāras*), consciousness (*viññāṇa*), form and body (*nāma-rūpa*), the five sense organs and the mind (*saḷāyatana*), contact (*phassa*), feeling-response (*vedanā*), craving (*taṇhā*), grasping for an object (*upādāna*), action toward life (*bhava*), birth (*jāti*), and old age and death (*jarāmaraṇa*). Thus, the misery that is bound up with all sensate existence is accounted for by a methodical chain of causation.

The chain of causation

The law of dependent origination of the various aspects of becoming remains invariable and fundamental in all schools of Buddhism. There are, however, diverse interpretations.

The Eightfold Path. Given the awareness of this law, the question arises as to how one may escape the continually renewed cycle of birth, suffering, and death. Here ethical conduct enters in. It is not enough to know that misery pervades all existence and to know the way in which life evolves; there must also be a purification that leads to the overcoming of this process. Such a liberating purification is effected by following the Noble Eightfold Path constituted by right views, right aspirations, right speech, right conduct, right livelihood, right effort, right mindfulness, and right meditational attainment. The term right (true or correct) is used to distinguish sharply between the precepts of the Buddha and other teachings.

Nirvana. The aim of religious practice is to be rid of the delusion of ego, thus freeing oneself from the fetters of this mundane world. One who is successful in doing so is said to have overcome the round of rebirths and to have achieved enlightenment. This is the final goal—not a paradise or a heavenly world.

The living process is likened to a fire burning. Its remedy is the extinction of the fire of illusion, passions, and cravings. The Buddha, the Enlightened One, is one who is no longer kindled or enflamed. Many poetic terms are used to describe the state of the enlightened human being—the harbour of refuge, the cool cave, the place of bliss, the farther shore. The term that has become famous in the West is nirvana, translated as dying out—that is, the dying out

The dying out

in the heart of the fierce fires of lust, anger, and delusion. But nirvana is not extinction, and indeed the craving for annihilation or nonexistence was expressly repudiated by the Buddha. Buddhists search not for mere cessation but for salvation. Though nirvana is often presented negatively as "release from suffering," it is more accurate to describe it in a more positive fashion: as an ultimate goal to be sought and cherished.

The Buddha left indeterminate questions regarding the destiny of persons who have reached this ultimate goal. He even refused to speculate as to whether such purified saints, after death, continued to exist or ceased to exist. Such questions, he maintained, were not relevant to the practice of the path and could not in any event be answered from within the confines of ordinary human existence.

Though it is true that the Buddha avoided discussion of the ultimate condition that lay beyond the categories of the phenomenal world, he often affirmed the reality of the religious goal. For example, he is reported to have said: "There is an unborn, an unoriginated, an unmade, an uncompounded; were there not, there would be no escape from the world of the born, the originated, the made, and the compounded."

In his teaching, the Buddha strongly asserted that the ontological status and character of the unconditioned nirvana cannot be delineated in a way that does not distort or misrepresent it. But what is more important is that he asserted with even more insistence that nirvana can be experienced—and experienced in this present existence—by those who, knowing the Buddhist truth, practice the Buddhist path. (Gi.T./H.Na./F.E.R.)

Historical development

INDIA

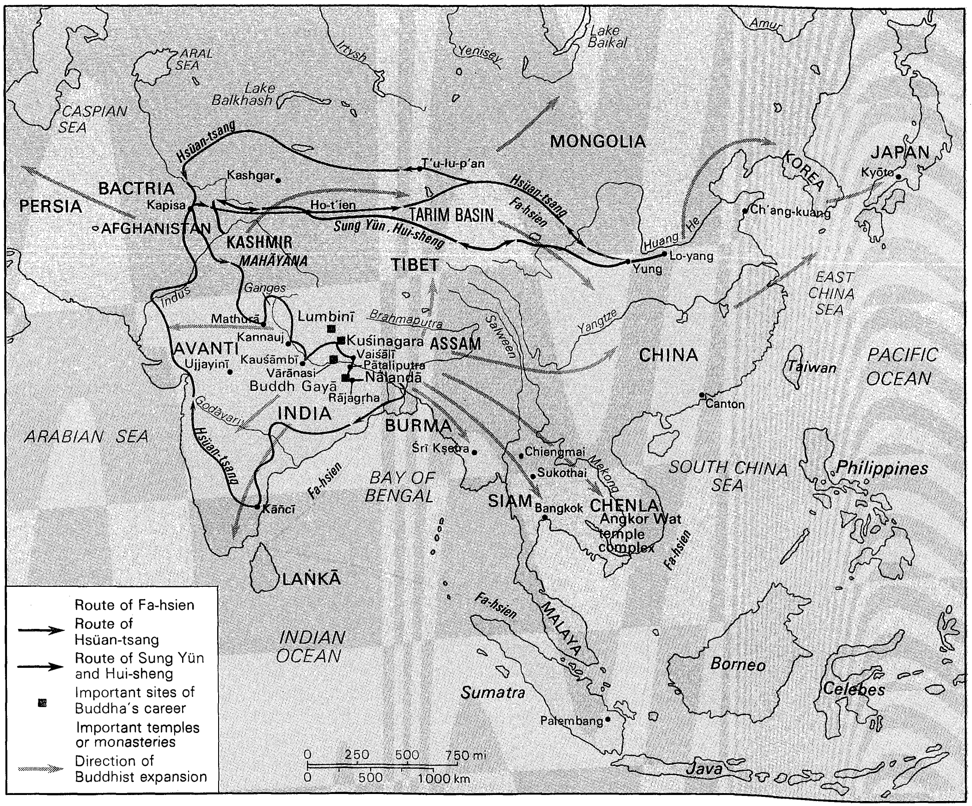
Expansion of Buddhism. The Buddha was a charismatic leader who discovered and proclaimed a religious message and founded a distinctive religious community. Some of the members of that community were, like the Buddha himself, wandering ascetics. Others were laypersons who venerated the Buddha, followed those aspects of his teachings that were relevant to them, and provided the wandering ascetics with the material support that they required.

During the first several centuries after the Buddha's death, the story of his life was remembered and embellished, his teachings were preserved and developed, and the community that he had established became a significant religious force. Many of the followers of the Buddha who were wandering ascetics began to settle in permanent monastic establishments and to develop the procedures needed to maintain large monastic institutions. At the same time, the Buddhist laity came to include important members of the economic and political elite.

During the first century of its existence Buddhism spread from its place of origin in Magadha and Kosala throughout much of northern India, including the areas of Mathurā (modern Madurai) and Ujjayanī in the west. According to the Buddhist tradition, invitations to the Council of Vesālī (Sanskrit: Vaiśālī), held just over a century after the Buddha's death, were sent to monks living in many distant places in northern and central India. By the middle of the 3rd century BC, Buddhism had gained the favour of a Mauryan king who had established an empire that extended from the Himalayas in the north almost as far south as Sri Lanka.

The Mauryan dynasty

To the rulers of the kingdoms and republics arising in northeastern India, the patronage of heteroprax sects (those with differing practices) was one way of counterbalancing the enormous political power enjoyed by Brahmans (high-caste Hindus) in the affairs of state. The first Mauryan emperor, Candra Gupta (*c.* 321–*c.* 297 BC), patronized Jainism and finally became a Jaina monk. His grandson, Aśoka, who ruled over the greater part of the subcontinent from about 270 to 230 BC, became the archetypal Buddhist king. Aśoka attempted to establish in his realm a "true *dhamma*" based on the virtues of self-control, impartiality, cheerfulness, truthfulness, and goodness. Though he did not found a state church, he did attempt to forge a Buddhist-oriented religiopolitical culture that would include Hindu, Jaina, Ājīvika (Ājīvaka), and Buddhist alike. His aim was to create a religious and social milieu that would enable all "children of the king" to live happily and attain heaven in the next life. Thus, he created a "welfare state" by setting up medical assistance for men and beasts, maintaining reservoirs and canals, and promoting trade. A system of *dhamma* officers (*dhamma-*



Important sites and routes of expansion of early Buddhism.

mahāmattas) was set up to provide for the empire magistrates, district attorneys, preachers, bureaucrats, social workers, and spies. The lay ethic preached by the king of the *dhamma* (*dhamma-rāja*) and his officers was focused on the layman's obligations in this world. Though Aśoka created a new ideal of kingship that would have powerful repercussions throughout the later Buddhist world, the various problems posed by a state of such vast dimensions in India proved greater than he could solve. Soon after Aśoka's death, the Mauryan empire began to crumble.

Although Buddhists seem to have suffered some persecutions during the subsequent Śuṅga–Kāṇva period (185–28 BC), Buddhism succeeded in maintaining and even expanding its influence. Buddhist monastic centres and magnificent Buddhist monuments such as the great stupas at Bhārhut and Sāñchi were established throughout the subcontinent, and these institutions often received royal patronage. In the early centuries of the Common era, Buddhism was especially flourishing in northwestern India, and from there it spread rapidly into Central Asia and China.

Buddhism under the Guptas and Pālas. By the time of the Gupta dynasty (*c.* AD 320–*c.* 600), Buddhism in India was being affected by the revival of Brahmanic religion and the rising tide of bhakti (a devotional movement that emphasized the intense love of a devotee for a personal god). During this period, for example, some Hindus were practicing devotion to the Buddha, whom they regarded as an avatar (incarnation) of the Hindu deity Vishnu.

During the Gupta period some monasteries joined together to form monastic centres (*mahāvihāras*) that functioned as universities. The most famous of these, located at Nālandā, had a curriculum that went far beyond the bounds of traditional Buddhism. Nālandā soon became the leading centre for the study of Mahāyāna, which was rapidly becoming the dominant Buddhist tradition in India.

Visits of Chinese pilgrims

Though Buddhist institutions seemed to be faring well under the Guptas, various Chinese pilgrims visiting India between AD 400 and 700 could discern an internal decline in the Buddhist community and the beginning of the reabsorption of Indian Buddhism by Hinduism. Among these pilgrims were Fa-hsien, Sung Yün, Hui-sheng, Hsüan-tsang, and I-ching.

The accounts of these Chinese travelers provide invaluable information about Asian cultures from the Sāsānian (Persian) empire in the west to Sumatra and Java in the east, and from Turfan in Central Asia to Kāñchi in the south of India. In 399 Fa-hsien left China, crossed the Gobi (Desert), and visited various holy places in India. He then returned to China via Sri Lanka and Java, taking with him numerous Buddhist scriptures and statues. The most famous of the Chinese travelers was the 7th-century monk Hsüan-tsang. When he arrived in northwestern India, he found "millions of monasteries" reduced to ruins by the Huns, a nomadic Central Asian people. Many of the remaining Buddhists were developing their own form of Tantrism, an esoteric psychic-physical system of belief and practice. In the northeast, Hsüan-tsang visited various holy places and studied Yogācāra, a Mahāyāna system, and Indian philosophy at Nālandā. After visiting Assam and southern India he returned to China with some 600 sutras.

After the destruction of numerous Buddhist monasteries in the 6th century AD by the Huns, Buddhism revived, especially in the northeast, where it flourished for a time under the Buddhist Pāla kings (8th–12th century AD). These kings continued to protect the great monastic establishments (*mahāvihāras*), building such new centres as Odantapurī, near Nālandā, and establishing a system of supervision for all such institutions. Under the Pālas, Tantric Buddhism (*i.e.,* Vajrayāna) became the dominant sect. Adepts of this sect, called siddhas, identified nirvana with the passions, maintaining that one could "touch the deathless element with his body." Though some of its practices seemed excessive, scholars of this school sought to revalorize some of the most archaic elements in Indian religion. During this period, the university of Nālandā became a centre for the study of Tantric Buddhism and the practice of Tantric magic and rituals. Under the Pāla kings, contacts with China decreased as Indians began to turn their attention to Tibet and Southeast Asia.

The decline of Buddhism in India. With the collapse of the Pāla dynasty in the 12th century, Buddhism suffered another defeat, and this time it did not recover. Though some pockets of Buddhist influence remained, the Buddhist presence in India became so negligible that it could hardly be noticed.

Scholars do not know all the factors that contributed to the demise of Buddhism in its original homeland. Some have maintained that Buddhism was so tolerant of other faiths that it was simply reabsorbed by a revitalized Hindu tradition. This did occur, although Indian Mahāyānists occasionally displayed a hostile attitude toward bhakti and toward Hinduism in general. However, there was another factor that was very important as well: Buddhism in India, having become mainly a monastic movement, probably paid little heed to the laity. Some monasteries became wealthy enough to have slaves and hired labourers to care for the monks and tend the lands they owned. Thus, after the Muslim invaders sacked the Indian monasteries in the 12th century AD, Buddhists had little basis for recovery. After the destruction of the monasteries, the Buddhist laity showed little interest in restoring the "Way."

Destruction of the monasteries

Contemporary revival. At the beginning of the 20th century Buddhism was virtually extinct in India. Since the early 1900s, however, a significant Buddhist presence has been reestablished. In the early decades of the 20th century a number of Buddhist societies were organized by Indian intellectuals who found in Buddhism an alternative to a Hindu tradition that they could no longer accept; an alternative that was, in addition, part of the cultural heritage of India. Following the Chinese conquest of Tibet in the late 1950s, there was an influx of Tibetan Buddhists who established a highly visible Buddhist community in northern India. In addition, the incorporation of Sikkim in 1975 into the Republic of India has brought into the modern Indian nation a small Himalayan society that has a strong Buddhist tradition related to the Vajrayāna Buddhism of Tibet.

The major component in the 20th-century resurgence of Buddhism in India has, however, been the mass conversion of large numbers of people from the so-called scheduled castes (formerly called Untouchables). This conversion movement, originally led by Bhimrao Ramji Ambedkar, began in the 1950s. In October 1956 Ambedkar and several hundred thousand of his followers converted to Buddhism, and—although accurate figures are difficult to determine—the group has continued to grow. Some estimates indicate that the number of converts is as high as four million. This group, which in the past has tended to favour the Theravāda version of Buddhism, is developing its own distinctive patterns of Buddhist teaching and practice.

SRI LANKA AND SOUTHEAST ASIA

The first clear evidence of the spread of Buddhism outside India dates from the reign of King Aśoka (3rd century BC). According to his inscriptions, Aśoka sent Buddhist emissaries not only to many different regions of the subcontinent but also into certain border areas as well. It is certain that Aśokan emissaries were sent to Sri Lanka and to an area called Suvarṇabhūmi that many modern scholars have identified with the Mon country in southern Myanmar (Burma) and central Thailand.

Transmission of Buddhism by Mauryan missionaries

Sri Lanka. According to the Sinhalese tradition, Buddhism took root in Sri Lanka with the arrival of Aśoka's son Mahinda and his six companions. Sent as missionaries by the Mauryan emperor, these travelers converted King Devānampiya Tissa and many of the nobility. Under King Tissa, the Mahāvihāra monastery was built, an institution that was to become the centre of Sinhalese orthodoxy. After Tissa's death (*c.* 207 BC) Sri Lanka fell into the hands of the South Indians until the time of Duṭṭhagāmaṇī (101–77 BC), a descendant of Tissa, who overthrew King Eḷāra. During this time, as a reaction to the threat posed by the South Indians, Buddhism and Sri Lankan political formations became closely intertwined. Again, it was probably because of this danger that the Pāli canon was first written down under King Vaṭṭagāmaṇī Abhaya in the 1st century BC. This king also built the Abhayagiri monastery, the main centre of the various Mahāyāna movements in Sri Lanka. These heterodox tendencies were openly supported by King Mahāsena (AD 276–303). Under Mahāsena's son, Śrī Meghavaṇṇa, the "Tooth of the Buddha" was brought to Abhayagiri and made the national palladium.

During the 1st millennium AD in Sri Lanka, the ancient Theravāda tradition coexisted with various forms of Hinduism, Mahāyāna Buddhism, and Buddhist Tantrism. Beginning in the 10th century—as Buddhism was declining in India—Sri Lanka became a major locus of a Theravāda Buddhist revival. As a result of this revival, Sri Lanka became a Theravāda kingdom, with a sangha that was unified under Theravāda auspices and a monarch who legitimated his rule in Theravāda terms. The new Theravāda tradition that was established spread from Sri Lanka into Southeast Asia, where it exerted a powerful influence.

In modern times Sri Lanka fell prey to the Western colonial powers (to the Portuguese in 1505–1658, the Dutch in 1658–1796, and finally the British in 1796–1947). Under King Kittisiri Rājasiha (1747–81) the ordination lineage was once again renewed, this time by monks recruited from Thailand.

The monastic community in Sri Lanka is now divided into three major bodies: (1) the Siam Nikaya, founded in the 18th century, a conservative and wealthy sect that admits only members of the Goyigama, the highest Sinhalese caste, (2) the Amarapura sect, founded in the 19th century, which has opened its ranks to members of lower castes, and (3) the reformed splinter group from the Siam Nikaya called the Ramanya sect. Among the laity several reform groups have been established. Among these the Sarvodaya community headed by A.T. Ariyaratne is especially important. This group has established religious, economic, and social development programs that have had a significant impact on Sinhalese village life.

Since Sri Lanka attained its independence from the British in 1947, the country has been increasingly drawn into a conflict between the Sinhalese Buddhist majority and the Tamil Hindu minority. Many Sinhalese Buddhists, including some monks, have closely associated their religion with the political agenda of the more militant Sinhalese nationalists. A few Buddhist leaders have, however, tried to adopt a more moderate position and to encourage a negotiated solution, the traditional basis for Sri Lankan political formation.

Southeast Asia. The peoples of Southeast Asia have not been mere satellites of the more powerful Indian and Chinese civilizations. On the contrary, the cultures that arose in these three vast areas might better be thought of as alternative developments that occurred within a greater Austroasiatic civilization, sometimes called "Asia of the monsoons." Therefore, the transmission of Buddhism and Hinduism to Southeast Asia can be regarded as the spread of the religious symbols of the more "advanced" elements within this Austroasiatic cluster to peoples sharing some of the basic religious presuppositions and traditions.

The main sources of Buddhist influence

In Southeast Asia the Buddhist impact has been made in very different ways in three different regions. In two of these (the region of Malaysia/Indonesia and the region on the mainland extending from Myanmar to southern Vietnam), the main connections have been via trade routes with India and Sri Lanka. In Vietnam the main connections have been with China.

Malaysia and Indonesia. Though some scholars locate the Suvarṇabhūmi ("Land of Gold"), to which Aśokan missionaries were supposedly sent, somewhere on the Malay Peninsula or in Indonesia, this is probably not accurate. It is, however, quite certain that Buddhism reached these areas by the beginning centuries of the 1st millennium AD.

With the help of Indian missionaries such as the monk Guṇavarman, Buddhism had gained a firm foothold on Java well before the 5th century AD. Buddhism was also introduced at about this time in Sumatra, and, by the 7th century, the king of Śrīvijaya on the island of Sumatra was a Buddhist. When the Chinese traveler I-ching visited this kingdom in the 7th century, he noted that Hīnayāna was dominant in the area but that there were also a few Mahāyānists. It was also in the 7th century that the great scholar Dharmapāla from Nālandā visited Indonesia.

The Śailendra dynasty, which ruled over the Malay Peninsula and a large section of Indonesia from the 7th to the 9th century, promoted the Mahāyāna and Tantric forms of Buddhism. During this period major Buddhist monuments were erected in Java, among them the marvelous Borobuḍur, which is perhaps the most magnificent of all Buddhist stupas. From the 7th century onward, Vajrayāna Buddhism spread rapidly throughout the area. King Kertanagara of Java (reigned 1268–92) was especially devoted to Tantric practice.

In the Malay Peninsula and Indonesia, as in India, Buddhism gradually lost its hold during the first half of the 2nd millennium AD. In many areas Buddhism was assimilated to Hinduism, forming a Hindu-oriented amalgam that in some places (for example in Bali) has persisted to the present. In most of Malaysia and Indonesia, however, both Hinduism and Buddhism were replaced by Islām, which remains the dominant religion in the area. (In modern Indonesia and Malaysia, Buddhism exists as a living religion only among the Chinese minority, but there is a growing community of converts, with its greatest strength in the vicinity of Borobuḍur.)

Suvarṇabhūmi

From Myanmar to the Mekong delta. A second pattern of Buddhist expansion in Southeast Asia developed in the mainland area that extends from Myanmar in the north and west to the Mekong delta in the south and east. According to the local Mon/Burman traditions, this is the area of Suvarṇabhūmi that was visited by missionaries from the Aśokan court. It is known that, by the early centuries of the 1st millennium AD, Buddhist kingdoms were beginning to appear in this region. In Myanmar and Thailand—despite the presence of Hindu, Mahāyāna, and Vajrayāna elements—the more conservative Hīnayāna forms of Buddhism were especially prominent throughout the 1st millennium AD. Farther to the east and south, in what is now Kampuchea (Cambodia) and southern Vietnam, various combinations of Hinduism, Mahāyāna Buddhism, and Vajrayāna Buddhism became dominant. Throughout much of the history of Angkor, the great imperial centre that dominated Kampuchea and much of the surrounding areas for many centuries, Hinduism seems to have been the preferred tradition, at least among the elite. In the late 12th and early 13th centuries, however, the Buddhist King Jayavarman VII built a new capital called Angkor Thom that was dominated by Mahāyāna/Vajrayāna monuments; these monuments represent one of the high points of Buddhist architectural achievement.

In mainland Southeast Asia, as in Sri Lanka, a Theravāda reform movement began to develop in the 11th century. Drawing heavily on the Theravāda heritage that had been preserved among the Mon in southern Myanmar, as well as on the new reform tradition that was developing in Sri Lanka, this revival soon established the Theravāda tradition as the most dynamic tradition in Myanmar, where the Burmans had conquered the Mon. By the late 13th century the reform movement had spread to Thailand, where the Thai were gradually displacing the Mon as the dominant population. Within another two centuries the Theravāda reformers had spread their tradition to Kampuchea and Laos.

The Theravāda preeminence that was thus established remained basically intact throughout the area during the remainder of the premodern period. The arrival of the Western powers in the 19th century, however, brought important changes. In Thailand, which retained its independence, a process of gradual reform and modernization took place. During the 19th century leadership in the reform and modernization process was taken by a new Buddhist sect, the Thammayut Nikāya, which was established and supported by the reigning Chakri dynasty. More recently, the reform and modernization process has become more diversified and has affected virtually all segments of the Thai Buddhist community.

Two Buddhist groups, Santi Asoke (founded 1975) and Dharmakaya, are especially interesting. Because of their hard-line demands for religious and moral reforms, both groups are at odds with the established ecclesiastical hierarchy. But, despite pressures from the government, they have acquired a large popular following.

In the other Theravāda countries in Southeast Asia, Buddhism has had a much more difficult time. In Myanmar, which endured an extended period of British rule, the sangha and the structures of Buddhist society have been seriously disrupted. Under the military regime of General Ne Win, established in 1962, reform and modernization were limited in all areas of national life, including religion. In Laos and Kampuchea, both of which suffered an extended period of French rule followed by the devastation of the Vietnam War and the violent imposition of communist rule, the Buddhist community has been severely crippled. During the late 20th century, however, many signs of a Buddhist revival began to appear.

Vietnam. There are some indications that Vietnam was involved in the early sea trade between India, Southeast Asia, and China and that Buddhism reached the country around the beginning of the 1st millennium AD, brought by missionaries traveling between India and the Chinese empire. The northern part of what is now Vietnam had been conquered by the Chinese empire in 111 BC; it remained under Chinese rule until AD 939. In the south there were two Indianized states, Funan (founded during the 1st century AD) and Champa (founded AD 192). In these areas both Hīnayāna and Mahāyāna traditions were represented. The traditions that most affected the long-term development of Buddhism in Vietnam, however, were Zen and Pure Land traditions introduced from China into the northern and central sections of the country beginning in the 6th century AD.

Funan and
Champa

The first *dhyāna* (Zen; Vietnamese: *thiên*), or "meditation," school was introduced by Vinītaruci, an Indian monk who had come to Vietnam from China in the 6th century. In the 9th century a school of "wall meditation" was introduced by the Chinese monk Vo Ngon Thong. A third major Zen school was established in the 11th century by the Chinese monk Thao Durong. From 1414 to 1428 Buddhism in Vietnam was persecuted by the Chinese, who had again conquered the country. Tantrism, Taoism, and Confucianism were also filtering into Vietnam at this time. Even after the Chinese had been driven back, a Chinese-like bureaucracy closely supervised the Vietnamese monasteries. The clergy was divided between the highborn and Sinicized (Chinese-influenced), on the one hand, and those in the lower ranks who often were active in peasant uprisings.

During the modern period these Mahāyāna traditions centred in northern and central Vietnam have coexisted with Theravāda traditions that have spilled over from Kampuchea in the south. Rather loosely joined together, the Vietnamese Buddhists managed to preserve their traditions through the period of French colonial rule in the 19th and 20th centuries. During the struggle between North and South Vietnam in the 1960s and early '70s, many Buddhists worked to achieve peace and reconciliation, but they met with little success. Under the communist regime that completed its victory in Vietnam in the early 1970s, conditions have been difficult, but Buddhism has persisted. Reports in the late 1980s and early '90s indicated that new signs of vitality were beginning to appear.

CENTRAL ASIA AND CHINA

Central Asia. The spread of Buddhism into Central Asia is still not completely understood by historians. But, however murky the details may be, it is clear that the trade routes that ran from northwestern India to northern China facilitated both the introduction of Buddhism and the maintenance, for many centuries, of a flourishing Buddhist culture.

By the beginning of the Common era, Buddhism had probably been introduced into eastern Turkistan. According to tradition, a son of Aśoka founded the kingdom of Khotan around 240 BC. The grandson of this king supposedly introduced Buddhism to Khotan, where it became the state religion. On more secure historical grounds, it is clear that the support given by the Indo-Scythian king Kaniṣka of the Kushān (Kuṣāṇa) dynasty, which ruled in northern India, Afghanistan, and parts of Central Asia in the 1st to 2nd century AD, encouraged the spread of Buddhism into Central Asia. Kaniṣka purportedly called an important Buddhist council; he patronized the Gandhāra school of Buddhist art, which introduced Greek and Persian elements into Buddhist iconography; and he supported Buddhist expansion within a vast region that extended far into the Central Asian heartland. In the northern part of Chinese Turkistan, Buddhism spread from Kucha (K'u-ch'e) to the kingdoms of Agnideśa (Karashahr), Kao-ch'ang (Turfan), and Bharuka (Aksu). According to Chinese travelers who visited Central Asia, the Hīnayānists (at least at the time of their visits) were strongest in Turfan, Shanshan, Kashgar, and Kucha, while Mahāyāna strongholds were located in Yarkand and Khotan.

Expansion
of
Buddhism

In Central Asia there was a confusing welter of languages, religions, and cultures, and, as Buddhism interacted with these various traditions, it changed and developed. Shamanism, Zoroastrianism, Nestorian Christianity, and Islām all penetrated these lands and coexisted with Buddhism. For example, some of the Mahāyāna bodhisattvas, such as Amitābha, may have been inspired, in part, by Zoroastrian influence. There is also evidence of some degree of syncretism between Buddhism and Manichaeism, an Iranian dualistic religion that was founded in the 3rd century AD.

Buddhism continued to flourish in parts of Central Asia until the 11th century, particularly under the patronage of the Uighur Turks. With the increasingly successful incursions of Islām (beginning in the 7th century AD) and the decline of the T'ang dynasty (618–907) in China, however, Central Asia ceased to be the important crossroads of Indian and Chinese culture that it once had been. Buddhism in the area gradually became a thing of the past.

China. Although there are reports of Buddhists in China as early as the 3rd century BC, Buddhism was not actively propagated in that country until the early centuries of the Common era. Tradition has it that Buddhism was introduced after the Han emperor Ming Ti (reigned AD 57/58–75/76) had a dream of a flying golden deity that was interpreted as a vision of the Buddha. Accordingly, the emperor dispatched emissaries to India who subsequently returned to China with the *Sutra in Forty-two Sections,* which was deposited in a temple outside the capital of Lo-yang. In actuality, Buddhism entered China gradually, first primarily through Central Asia and, later, by way of the trade routes around and through Southeast Asia.

The early centuries. The Buddhism that first became popular in China during the Han dynasty was deeply coloured with magical practices, making it compatible with popular Chinese Taoism (a combination of folk beliefs and practices and philosophy). Instead of the doctrine of no-self, early Chinese Buddhists taught the indestructibility of the soul. Nirvana became a kind of immortality. They also taught the theory of karma, the values of charity and compassion, and the need to suppress the passions. Until the end of the Han dynasty, there was a virtual symbiosis between Taoism and Buddhism and a common propagation of the means for attaining immortality through various ascetic practices. It was widely believed that Lao-tzu, the founder of Taoism, had been reborn in India as the Buddha. Many Chinese emperors worshiped Lao-tzu and the Buddha on the same altar. The first translations of Buddhist sutras into Chinese—namely those dealing with such topics as breath control and mystical concentration—utilized a Taoist vocabulary to make the Buddhist faith intelligible to the Chinese.

After the Han period, in the north of China, Buddhist monks were often used by non-Chinese emperors for their political-military counsel as well as for their skill in magic. At the same time, in the south, Buddhism began to penetrate the philosophical and literary circles of the gentry. One of the most important contributions to the growth of Buddhism in China during this period was the work of translation. The most important of the early translators was the learned monk Kumārajīva, who, before he was brought to the Chinese court in AD 401, had studied the Hindu Vedas, the occult sciences, and astronomy, as well as the Hinayāna and Mahāyāna sutras.

The translations of
Kumārajīva

During the 5th and 6th centuries AD Buddhist schools from India became established, and new, specifically Chinese schools began to form. Buddhism was becoming a powerful intellectual force in China, monastic establishments were proliferating, and Buddhism was becoming well-established among the peasantry. Thus, it is not surprising that, when the Sui dynasty (581–618) established its rule over a reunified China, Buddhism flourished as a state religion.

Developments during the T'ang dynasty (618–907). The golden age of Buddhism in China occurred during the T'ang dynasty. Though the T'ang emperors were usually Taoists themselves, they tended to favour Buddhism, which had become extremely popular. Under the T'ang the government extended its control over the monasteries and the ordination and legal status of monks. From this time forward, the Chinese monk styled himself simply *ch'en,* or "a subject."

During this period several Chinese schools developed their own distinctive approaches. Some of them produced comprehensive systematizations of the vast body of Buddhist texts and teachings. There was a great expansion in the number of Buddhist monasteries and the amount of land they owned. It was also during this period that many scholars made pilgrimages to India, heroic journeys that greatly enriched Buddhism in China, both by the texts that were acquired and by the intellectual and spiritual inspiration that was brought from India. Buddhism was never able to replace its Taoist and Confucian rivals, however, and in 845 the emperor Wu-tsung began a major persecution. According to records, 4,600 Buddhist temples and 40,000 shrines were destroyed, and 260,500 monks and nuns were forced to return to lay life.

Buddhism after the T'ang. Buddhism in China never recovered completely from the great persecution of 845. It did maintain much of its heritage, however, and it continued to play a significant role in the religious life of China. On the one hand, Buddhism retained its identity as Buddhism and generated new forms through which it was expressed. These included texts such as the *yü lu,* or "recorded sayings," of famous teachers that were oriented primarily toward monks, as well as more literary creations such as the *Journey to the West* (written in the 16th century) and *The Dream of the Red Chamber* (18th century). On the other hand, Buddhism coalesced with the Confucian–Neo-Confucian and Taoist traditions to form a complex multi-religious ethos within which all three traditions were more or less comfortably encompassed.

The Ch'an and Pure Land schools

Among the various schools the two that retained the greatest vitality were the Ch'an school (better known in the West by its Japanese name, Zen) which was noted for its emphasis on meditation, and the Pure Land tradition, which emphasized Buddhist devotion. The former school exerted the greatest influence among the cultured elite. It did so through various media, including the arts. For example, Ch'an artists during the Sung dynasty (960–1279) had a decisive impact on Chinese landscape painting. Artists used images of flowers, rivers, and trees, executed with sudden, deft strokes, to evoke an insight into the flux and emptiness of all reality. The Pure Land tradition exerted a greater influence on the population as a whole and was sometimes associated with secret societies and peasant uprisings. But the two seemingly disparate traditions were often very closely linked. In addition, they were mixed with other Buddhist elements such as the so-called "masses for the dead" that had originally been popularized by the practitioners of Esoteric Buddhism.

During the early decades of the 20th century, China experienced a Buddhist reform movement aimed at revitalizing the Chinese Buddhist tradition and adapting Buddhist teachings and institutions to modern conditions. However, the disruptions caused by the Sino-Japanese War and the subsequent establishment of a communist government have not been helpful to the Buddhist cause. The Buddhist community was the victim of severe repression during the Cultural Revolution (1966–69). Since 1976 the Chinese government has pursued a more tolerant policy, but the extent of continuing Buddhist vitality is difficult to determine.

KOREA AND JAPAN

Korea. Buddhism was first introduced into the Korean region when it was divided into the three kingdoms of Paekche, Koguryŏ, and Silla. After Buddhism was brought to the northern kingdom of Koguryŏ from China in the 4th century, it gradually spread throughout the other Korean kingdoms. As often happened, the new faith was first accepted by the court and then extended to the people. After the unification of the country by the kingdom of Silla in the 660s, Buddhism began to flourish throughout Korea. The monk Wŏnhyo (617–686) was one of the most impressive scholars and reformers of his day. He was married and taught an "ecumenical" version of Buddhism that included all branches and sects. He tried to use music, literature, and dance to express the meaning of Buddhism. Another scholar of the Silla era was Ŭi-sang (625–702), who went to China and returned to spread the Hwaŏm (Hua-yen in Chinese) sect in Korea. The Chinese Ch'an sect (Zen) was introduced in the 8th century and, by absorbing the Korean versions of Hua-yen, T'ien-t'ai, and Pure Land, gradually became the dominant school of Buddhism in Korea, as it did in Vietnam.

Early Korean Buddhism was characterized by a this-worldly attitude. It emphasized the pragmatic, nationalistic, and aristocratic aspects of the faith. Still, an indigenous tradition of shamanism influenced the development of popular Buddhism throughout the centuries. Buddhist monks danced, sang, and performed the rituals of shamans.

The Koryŏ period

During the Koryŏ period (935–1392), Korean Buddhism reached its zenith. During the first part of this period the Korean Buddhist community was active in the publication of the *Tripitaka Koreana,* one of the most inclusive editions of the Buddhist sutras up to that time. After 25 years of research, a monk by the name of Ŭich'ŏn (1055–1101) published an outstanding three-volume bibliography of Buddhist literature. Ŭich'ŏn also sponsored the growth of the T'ien-t'ai sect in Korea. He emphasized the need for cooperation between Ch'an and the other "Teaching" schools of Korean Buddhism.

Toward the end of the Koryŏ period, Buddhism began to suffer from internal corruption and external persecution, especially that promoted by the Neo-Confucians. The government began to put limits on the privileges of the monks, and Confucianism replaced Buddhism as the religion of the state. Though the Yi dynasty (1392–1910) continued these restrictions, Buddhist monks and laymen fought bravely against the invasion of the Japanese armies under Toyotomi Hideyoshi (1537–98) in 1592 and 1597. In the decade before the annexation of Korea by Japan (1910), some effort was made to unify Korean Buddhism. These efforts, as well as the subsequent efforts of Buddhist "missionaries" from Japan, were largely in vain.

Since the end of World War II, Buddhism in Korea has been hampered by communist rule in North Korea and by the great vitality of Christianity in South Korea. Despite these challenges, Buddhists, particularly in South Korea, have preserved the old traditions and initiated new movements.

Japan. *Introduction of Buddhism to Japan.* While Buddhism in China sent its roots down into the subsoil of the family system, in Japan it found anchorage in the nation itself. The Buddhism that was initially introduced into Japan in the 6th century from Korea was regarded as a talisman (charm) for the protection of the country. The new religion was accepted by the powerful Soga clan but was rejected by others, thus causing controversies that resembled the divisions caused by the introduction of Buddhism in Tibet. In both countries, some believed that the introduction of Buddhist statues had been an insult to the native deities, resulting in plagues and natural disasters. Only gradually were such feelings overcome. Though the Buddhism of the Soga clan was largely magical, under the influence of Prince Shōtoku, who became regent of the nation in 593, other aspects of Buddhism were emphasized. Shōtoku lectured on various scriptures that emphasized the ideals of the layman and monarch, and he composed a "Seventeen-Article Constitution" in which Buddhism was adroitly mixed with Confucianism as the spiritual foundation of the state. In later times he was widely regarded as an incarnation of the bodhisattva Avalokiteśvara.

Nara and Heian periods. During the Nara period (710–784), Buddhism became the state religion of Japan. Emperor Shōmu actively propagated the faith, making the imperial capital, Nara—with its "Great Buddha" statue (Daibutsu)—the national cult centre. Buddhist schools imported from China became established in Nara, and state-subsidized provincial temples (*kokubunji*) made the system effective at the local level.

The Tendai and Shingon schools

After the capital was moved to Heian-kyō (modern Kyōto) in 794, Buddhism continued to prosper. Chinese influence continued to play an important role, particularly through the introduction of new Chinese schools that became dominant at the royal court. Mount Hiei and Mount Kōya became the centres for the new T'ien-tai (Tendai) and Esoteric (Shingon) schools of Buddhism, which were characterized by highly sophisticated philosophies and complex and refined liturgies. Moreover, Buddhism interacted with the indigenous Shintō and local tradition, and various distinctively Japanese patterns of Buddhist-oriented folk religion became very popular.

New schools of the Kamakura period. The 12th and 13th centuries marked a turning point in Japanese history and in the history of Japanese Buddhism in particular. Late in the 12th century the imperial regime with its centre at Heian collapsed, and a new feudal government, or shogunate, established its headquarters at Kamakura. As a part of the same process, a number of new Buddhist leaders emerged and established schools of Japanese Buddhism. These reformers included proponents of the Zen traditions such as Eisai and Dōgen; Pure Land advocates such as Hōnen, Shinran, and Ippen; and Nichiren, the founder of a new school that gained considerable popularity. The distinctively Japanese traditions these creative reformers and founders established became—along with many very diverse synthetic expressions of Buddhist-Shintō piety—integral components of a Buddhist-oriented ethos that structured Japanese religious life into the 19th century. Also during this period many Buddhist groups allowed their clergy to marry, with the result that temples often fell under the control of particular families.

The premodern period to the present. Under the Tokugawa shogunate (1603–1867), Buddhism became an arm of the government. Temples were used for registering the populace; this was one way of preventing the spread of Christianity, which the feudal government regarded as a political menace. This association with the Tokugawa regime made Buddhism quite unpopular at the beginning of the Meiji period (1868–1912), at least among the elite. At that time, in order to set up Shintō as the new state religion, it was necessary for Japan's new ruling oligarchy to separate Shintō from Buddhism. This led to the confiscation of temple lands and the defrocking of many Buddhist priests.

During the period of ultranationalism (*c.* 1930–45), Buddhist thinkers called for uniting the East in one great "Buddhaland" under the tutelage of Japan. After the war, however, Buddhist groups, new and old alike, began to emphasize Buddhism as a religion of peace and brotherhood. During the postwar period the greatest visible activity among Buddhists has been among the "New Religions" such as Sōka-gakkai ("Value Creation Society") and Risshō-Kōsei-kai ("Society for Establishing Righteousness and Friendly Relations"). During this period, Sōka-gakkai entered politics with the same vigour it had traditionally shown in the conversion of individuals. Because of its highly ambiguous but conservative ideology, the Sōka-gakkai-based political party (the Kōmeitō) is regarded with suspicion and fear by many Japanese.

TIBET, MONGOLIA, AND THE HIMALAYAN KINGDOMS

Padmasambhava

Tibet. Buddhism, according to the Tibetan tradition, was first given recognition in Tibet during the reign of Srong-brtsan-sgam-po (*c.* 627–*c.* 650). This king had two queens who were early patrons of the religion and were later regarded in popular tradition as incarnations of the Buddhist savioruess Tārā. The religion received active encouragement from Khri-srong-lde-btsan, during whose reign (*c.* 755–797) the first Buddhist monastery in Tibet was built at Bsam-yas (Samye), the first seven monks were ordained, and the celebrated Indian Tantric master Padmasambhava was invited to Tibet. A great deal of legend surrounds Padmasambhava, who was a *mahāsiddha* ("master of miraculous powers"); he is credited with subduing the Bon spirits and demons (the spirits and demons associated with the indigenous religion of Tibet) and with subjugating them to the service of Buddhism. At the time, influences from Chinese Buddhism were strong, but it is recorded that at the Council of Bsam-yas (792–794) it was decided that the Indian tradition should prevail.

Following a period of suppression that lasted almost two centuries (from the early 800s to the early 1000s), Buddhism in Tibet enjoyed a revival. During the 11th and 12th centuries many Tibetans traveled to India to acquire and translate Buddhist texts and to receive training in Buddhist doctrine and practice. With the assistance of the renowned Indian master Atīśa, who arrived in Tibet in 1042, Buddhism became established as the dominant religion. From this point forward Buddhism was the primary culture of the elite, was a powerful force in the affairs of state, and penetrated deeply into all aspects of Tibetan life.

One of the great achievements of the Buddhist community in Tibet was the translation into Tibetan of a vast corpus of Buddhist literature, including the *Bka'-'gyur* ("Translation of the Buddha Word") and *Bstan-'gyur* ("Translation of Teachings") collections. The *Bka'-'gyur* contains six sections: (1) *Tantra,* (2) *Prajñāpāramitā,* (3) *Ratnakūṭa,* a collection of small Mahāyāna texts, (4) *Avataṃsaka,* (5) *Sūtra* (mostly Mahāyāna sutras, but some Hīnayāna texts are included), and (6) *Vinaya.* The *Bstan-'gyur* contains 224 volumes with 3,626 texts, divided into three major groups: (1) *stotra*s (hymns of praise) in one volume, including 64 texts, (2) commentaries on tantras in 86 volumes, including 3,055 texts, and (3) commentaries on sutras in 137 volumes, including 567 texts.

A major development in the history of Tibetan Buddhism occurred in the late 14th or early 15th century when a great Buddhist reformer named Tsong-kha-pa established the Dge-lugs-pa school, known more popularly as the Yellow Hats. In 1578, representatives of this school succeeded in converting the Mongol Altan Khan, and, under the Khan's sponsorship, their leader (the so-called third Dalai Lama) gained considerable monastic power. In the middle of the 17th century the Mongol overlords established the fifth Dalai Lama as the theocratic ruler of Tibet. The succeeding Dalai lamas, who were regarded as successive incarnations of the bodhisattva Avalokiteśvara, held this position during much of the remainder of the premodern period, ruling from the capital, Lhasa.

The fifth Dalai Lama instituted the high office of Panchen Lama for the abbot of the Tashilhunpo monastery, located to the west of Lhasa. The Panchen lamas were regarded as successive incarnations of the buddha Amitābha. Unlike the Dalai Lama, the Panchen Lama has usually been recognized only as a spiritual ruler.

Throughout much of Tibetan history many of the great monasteries were controlled by aristocratic abbots who were able to marry and pass along their monastic possessions to their sons. Monks were often warriors, and monasteries became armed fortresses. The Manchus in the 18th century and subsequently the British, the Nationalist Chinese, and the Chinese communists have all tried to exploit the division of power between the Panchen and the Dalai lamas for their own ends. In 1959, after the Dalai Lama fled to India, the Chinese communists took over his temporal powers.

Tibetan Buddhism under the Chinese

Under Chinese rule, Tibetan Buddhists have suffered periods of persecution, some of them severe. Not surprisingly, this has strengthened the bond between Buddhism and nationalist resistance.

Mongolia. The distinctive form of Buddhism that developed in Tibet has exerted a strong influence on neighbouring areas and peoples. Most important in this regard was the conversion of the Mongol tribes to the north and east of Tibet. There are some indications that Buddhism was present among the Mongols as early as the 4th century, but the sources are scarce. It is clear, however, that during the 13th century close relationships developed between the Mongol court in China and some of the leaders of Tibetan Buddhism. Kublai Khan himself became a supporter of the Tibetan form of Buddhism. Kublai Khan's Tibetan advisers helped to develop a block script for the Mongolian language, and many Buddhist texts were translated from Tibetan into Mongolian. In general, however, the religion failed to gain widespread popular support during this period.

In 1578 a new situation developed when the Altan Khan accepted the Dge-lugs-pa version of the Tibetan tradition and supported its spread among his followers at all levels of Mongol society. Over the centuries the Mongols developed their own very rich Buddhist traditions. Mongolian scholars translated a large corpus of texts from Tibetan, and they produced their own sophisticated original texts. The Mongols based their Buddhist doctrine, practice, and communal organization on Tibetan models, but they developed and adapted them in a distinctive way.

Between 1280 and 1368 China was part of the Mongol empire, and the Mongols established their variant of Tibetan Buddhism in China. When they no longer held power in China, they continued to maintain the traditions they had developed in their homeland in the Central Asian steppes. During the 20th century, however, Mongolian Buddhism has been undermined by the communist regimes that have ruled in the Mongol areas of the Soviet Union, Mongolia, and China.

The Himalayan kingdoms. Tibetan Buddhism has also exerted a considerable influence in the Himalayan kingdoms situated along Tibet's southern border. In Nepal Buddhism has been influenced by both India and Tibet. Though the Buddha was born in the southern part of the area that is now Nepal, at Lumbini, about 15 miles (24 kilometres) from Kapilavatthu (Kapilavastu), the Buddhist religion seems to have been actively propagated only later, probably under Aśoka. By the 8th century AD Nepal had fallen into the cultural orbit of Tibet. A few centuries later, as a result of the Muslim invasions of India, both Hindus (such as the Brahmanic Gurkha aristocracy) and Buddhists took refuge in the country. In modern times Buddhist prayer wheels and flags are reminders of the direct influence of Tibetan Buddhism. The Indian heritage is especially evident in the caste system that embraces Buddhists and non-Buddhists alike.

Bhutan

In Bhutan a Tibetan lama introduced Buddhism and a Tibetan style of hierarchical theocracy in the 17th century AD. The Buddhism practiced in Bhutan has been influenced by the Tibetan Bka'-brgyud-pa sect, which has stressed the magical benefits of living in caves and has not enforced on its clergy the discipline of celibacy. Buddhism in Bhutan, like Buddhism in Nepal, is coming into increasing contact with modernizing forces that are beginning to undermine many of its traditional practices.

BUDDHISM IN THE WEST

During the long course of Buddhist history, Buddhist influences have from time to time reached the Western world. Though the evidence is weak, some scholars have suggested that, about the beginning of the Christian era, Buddhist monks and teachings had reached as far as Egypt. There are occasional references to what seem to be Buddhist traditions in the writings of the Church Fathers. In addition, a version of the biography of the Buddha, known as the story of Barlaam and Josaphat, had a considerable dissemination in medieval Europe. In fact, the Buddha-figure in the story came to be recognized as a Christian saint.

Not until the modern period, however, is there evidence for a serious Buddhist presence in the Western world. The movement of Buddhism from Asia to the West that took place in the 19th and 20th centuries had two aspects. Beginning in the mid-19th century, Buddhism was introduced into the United States and other Western countries by large numbers of immigrants, first from China and Japan but more recently from other countries, especially countries of Southeast Asia. Buddhism gained a foothold among a significant number of Western intellectuals and—particularly during the 1960s and early '70s—among young people seeking new forms of religious experience and expression. The interest of Westerners in Buddhism has been increased by the work of Buddhist missionaries such as the Japanese scholar D.T. Suzuki (1870–1966) and a number of Tibetan Buddhist teachers who have come to the West since the Chinese conquest of their homeland in the late 1950s. (J.M.K./Gi.T./F.E.R.)

Sangha, society, and state

Buddhists of all times and places have recognized the importance of community life, and over the centuries there has developed a distinctive pattern involving a symbiotic relationship between monks (and in some cases nuns) and the lay community. The relationship between the monastics and the laity has differed from place to place and from time to time, but throughout most of Buddhist history both groups have played an essential role in the process of constituting and reconstituting the Buddhist world. Moreover, both the monastics and the laity have engaged in a variety of common and complementary religious practices that have expressed Buddhist orientations and values, structured Buddhist societies, and addressed the soteriological and practical concerns of Buddhist individuals.

The relationship between monastics and laypersons

MONASTIC INSTITUTIONS

The sangha is the assembly of Buddhist monks that has, from the origins of Buddhism, authoritatively studied, taught, and preserved the teachings of the Buddha. In their communities monks have served the laity through example and, as directed by the Buddha, through the teachings of morality (Pāli: *sīla;* Sanskrit: *śīla*). In exchange for their service the monks have received support from the laity, who thereby earn merit. Besides serving as the centre of Buddhist propaganda and learning, the monastery offers the monk an opportunity to live apart from worldly concerns, a situation that has usually been believed necessary or at least advisable in order to follow strictly the path that leads most directly to release.

The origin and development of the sangha. According to scholars of early Buddhism, at the time of the Buddha in northeastern India there existed numerous religious mendicants or almsmen who wandered and begged individually or in groups. These men had forsaken the life of a householder and the involvement with worldly affairs that this entails in order to seek a doctrine and form of practice which would meaningfully explain life and offer salvation. When such a seeker met someone who seemed to offer such a salvatory message, he would accept him as a teacher (guru) and wander with him. The situation of these mendicants is summed up in the greeting with which they met other religious wanderers. This greeting asked, "Under whose guidance have you accepted religious mendicancy? Who is your master (*sattha*)? Whose *dhamma* is agreeable to you?"

The groups of mendicants that had formed around a teacher broke their wanderings during the rainy season (*vassa*) from July through August. At this time they gathered at various rain retreats (*vassavāsa*), usually situated near villages. Here they would beg daily for their few needs and continue their spiritual quest. The Buddha and his followers may well have been the first group to found such a yearly rain retreat.

After the Buddha's death his followers did not separate but continued to wander and enjoy the rain retreat together. In their retreats the followers of the Buddha's teachings probably built their own huts and lived separately, but their sense of community with other Buddhists led them to gather fortnightly at the time of the full and new moons to recite the *Pātimokkha,* or declaration of their steadfastness in observing the monastic discipline. This ceremony, in which the laity also participated, was called the *uposatha.*

The *uposatha*

Within the first several centuries after the Buddha's death, the sangha came to include two different groups of monks. One retained the wandering mode of existence; this group has been a very creative force in Buddhist history and continues to play a role in contemporary

Buddhism, particularly in Sri Lanka and Southeast Asia. The other, much larger group gave up the life-in-the-forest and settled in permanent monastic settlements (viharas). There appear to be two major reasons for this change in the mode of living. First, the followers of the Buddha were able, through their confession of a common faith, to build up a certain coherent organization. Second, the laity gave meritorious gifts of land and raised buildings in which the followers of the Buddha might live permanently, assured of a supply of the staples of life and also fulfilling the Buddha's directive to minister to the laity. In this manner small viharas were raised in northeastern India and adjoining areas into which Buddhism spread. With the reign of King Aśoka in the 3rd century BC, further developments occurred. This king, who controlled much of India, took a protective interest in the unity and well-being of the Buddhist monastic community, and he promulgated a *dhamma* for his kingdom that shared a great deal with Buddhist teachings directed to the laity. As a result of his support and influence, Buddhism developed a more universal orientation.

In the post-Aśokan period, Buddhist monasteries grew in size and acquired a great deal of wealth. By about the 5th century AD there developed *mahāvihāra*s, or monastic centres, serving as universities, such as Nālandā. These universities were centres of Buddhist learning and propaganda, drawing monks from China and Tibet and sending forth missionaries to these lands. The institutions were open to the outside influence of a resurgent Hinduism, however, which is one of the factors that weakened Buddhism and led to its disappearance from India in the 13th century.

In all Buddhist countries, monasteries continued to serve as centres of missions and learning and as retreats. Different types of monastic establishments developed in particular areas and in particular contexts. In several regions there were at least two types of institutions. There were a few large public monasteries that usually functioned in greater or lesser accord with classical Buddhist norms. In addition, there were many smaller monasteries, often located in rural areas, that were much more loosely regulated. Often these were hereditary institutions in which the rights and privileges of the abbot were passed on to an adopted disciple. In areas where clerical marriage was practiced—for example, in medieval Sri Lanka and in post-Heian Japan—a tradition of blood inheritance developed.

Internal organization of the sangha. The development of the sangha from a group of wandering mendicants loosely bound together by their faith in the Buddha and his teachings, to monks living closely together in a permanent monastery necessitated the development of rules and a degree of hierarchical organization. It appears that the earliest organization within Indian monasteries was democratic in nature. This democratic nature arose from two important historical factors. First, the Buddha did not, as was the custom among the teachers of his time, designate a human successor. Instead, the Buddha taught that each monk should strive to follow the path that he had preached. This decision of the Buddha placed every monk on the same footing. There could be no absolute authority vested in one person, for the authority was the *dhamma* that the Buddha had taught. Second, the region in which Buddhism arose was noted for a system of tribal democracy, or republicanism. When a serious question demanded attention in the region, the male inhabitants would meet to decide upon a course of action, often electing a temporary ruler. This republican tradition, which supported the antiauthoritarian nature of the Buddha's teaching, was adopted by the early sangha.

The democratic nature of the early monasteries

When an issue arose, all the monks of the monastery assembled. The issue was put before the body of monks and discussed. If any solution was forthcoming, it had to be read three times, with silence signifying acceptance. If there was debate, a vote might be taken or the issue referred to committee or the arbitration of the elders of a neighbouring monastery. As the sangha developed, a certain division of labour and hierarchical administration was adopted. The abbot became the head of this administrative hierarchy and was vested with almost unlimited powers over monastic affairs. The antiauthoritarian character of Buddhism, however, continued to assert itself. In China, for instance, the abbot continued to refer all important questions to the assembled monks, who had elected him as their leader. Similarly, in Southeast Asian countries there has traditionally been a popular distaste for hierarchy, making rules difficult to enforce in the numerous almost independent monastic units.

As the Buddhist sangha developed, specific rules and rites were enacted that differ very little in all Buddhist monasteries even today. The rules by which the monks are judged and the punishments that should be assessed are found in the vinaya texts (vinaya literally means "that which leads"). The *Vinaya Piṭaka* of the Theravāda canon contains precepts that were supposedly given by the Buddha as he judged a particular situation. While in the majority of cases the Buddha's authorship can be doubted, the attempt is made to refer all authority to the Buddha and not to one of his disciples. The heart of the vinaya texts is the *Pātimokkha,* which, in the course of the sangha's development, became a list of monastic rules. The rules are recited by the assembled monks every fortnight, with a pause after each one so that any monk who has transgressed this rule may confess and receive his punishment. While the number of rules in the *Pātimokkha* differs in the various schools, with 227, 250, and 253, respectively in the Pāli, Chinese, and Tibetan canons, the rules are essentially the same. The first part of the *Pātimokkha* deals with the four gravest sins, which necessarily lead to expulsion from the monastery. They are sexual intercourse, theft, murder, and exaggeration of one's miraculous powers. The other rules, in seven sections, deal with transgressions of a lesser nature, such as drinking or lying.

The *Pātimokkha*

In the Theravāda countries—Sri Lanka, Myanmar, Thailand, Kampuchea, and Laos—the Buddhist monastic community is composed primarily of male monks and novices (the order of nuns died out in the Theravāda world more than a millennium ago, and contemporary efforts to reestablish it have met with only minimal success), white-robed ascetics (including various types of male and female practitioners who remain outside the sangha but follow a more or less renunciatory mode of life), and laymen and laywomen. In some Theravāda countries, notably in mainland Southeast Asia, boys or young men were traditionally expected to join the monastery for a period of instruction and meditation. Thus, the majority of men in these areas were (and to a lesser extent still are, at least in Myanmar and Thailand) directly involved with the monastic ethos. This practice has fostered a high degree of lay participation in monastic affairs.

In the Mahāyāna and Vajrayāna countries of China and Tibet there was traditionally a stage of one year before the aspirant could become a novice. This was a year of probation when the aspirant did not receive tonsure and remained subject to governmental taxation and service, while receiving instructions and performing menial tasks within the monastery. At the end of this one-year probationary period, the aspirant had to pass a test, including the recitation of part of a well-known sutra—the length depending upon whether the applicant was male or female—and a discussion on various doctrinal questions. In China, one usually did not progress beyond the novice stage unless he or she was of exceptional character or was affiliated with the government.

According to vinaya rules, entry into the sangha is an individual affair, dependent upon the wishes of the individual and his family. In some Buddhist countries, however, ordination was often under the control of the state, and the state conducted the examinations to determine entry or advancement in the sangha. In certain situations ordination could be obtained not only through such examinations but also by the favour of high officials or through the purchase of an ordination certificate from the government. This selling of ordination certificates was at times abused by the government in order to fill its treasury.

The life of a Buddhist monk was originally one of wandering, poverty, begging, and strict sexual abstinence. The monks were supposed to live only on alms, to wear clothes made from cloth taken from rubbish heaps, and to possess only three robes, one girdle, an alms bowl, a razor, a needle, and a water strainer used to filter insects from the drinking water (so as not to kill or imbibe them). Most Buddhist schools still stress celibacy, although some groups, particularly in Tibet and Japan, have relaxed the monastic discipline, and some Vajrayāna schools have allowed sexual intercourse as an esoteric ritual that contributes to the attainment of release. Begging, however, has tended in all schools to become merely a symbolic gesture used to teach humility or compassion or to raise funds for special purposes. Also, the growth of large monasteries has often led to compromises on the rule of poverty. While the monk might technically give up his property before entering the monastery—although even this rule is sometimes relaxed—the community of monks might inherit wealth and receive lavish gifts of land. This acquisition of wealth has led at times not only to a certain neglect of the Buddhist monastic ideal but also to the attainment of temporal power. This factor, in addition to the self-governing nature of Buddhist monasteries and the early Buddhist connection with Indian kingship, has influenced the interaction of the sangha and the state.

Relaxation of the poverty rule

SOCIETY AND STATE

Though Buddhism is sometimes described as a purely monastic, otherworldly religion, this is not accurate. In the earliest phases of the tradition the Buddha was pictured as a teacher who addressed not only renouncers but lay householders as well. Moreover, although he is not depicted in the early texts as a social reformer, he does address issues of social order and responsibility. (See, for example, the famous *Sigālovāda Sutta,* which has been called the "householder's vinaya.")

Throughout Buddhist history, Buddhists have put forth varying forms of social ethics based on notions of karmic justice (the "law" that good deeds will be rewarded with happy results while evil deeds will entail suffering for the one who does them); the cultivation of virtues such as self-giving, compassion, and evenhandedness; and the fulfillment of responsibilities to parents, teachers, rulers, and so on. Moreover, Buddhists have formulated various notions of cosmogony, cosmology, and soteriology that have provided legitimacy for the social hierarchies and political orders with which they have been associated. For the most part, Buddhism has played a conservative, moderating role in the social and political organization of various Asian societies, but the tradition has on occasion given rise to more radical and revolutionary movements as well.

Over the course of Buddhism's long history, the relationship between the Buddhist community and state authority has taken many forms. The early Buddhist sangha in India appears to have been treated by Indian rulers as a self-governing unit not subject to their power unless it proved subversive or was threatened by internal or external disruption. Aśoka, the Buddhist king whose personal faith and prestige helped Buddhism grow from a regional to a universal religion, appears to have been applying this policy of protection from disruption when he intervened in Buddhist monastic affairs to expel schismatics. He came to be remembered, however, as the Dharmarāja, the great king who protected and propagated the teachings of the Buddha.

In Theravāda countries Aśoka's image as a supporter and sponsor of the faith has traditionally been used to judge political authority. In general, Buddhism in Theravāda countries has been either heavily favoured or officially recognized by the government, so that the golden age in which there is a creative interaction between the government and the monks has been viewed as an obtainable goal. The sangha's role in this interaction has traditionally been to preserve the *dhamma* and to act as the spiritual guide and model, revealing to the secular power the need for furthering the welfare of the people. While the sangha and the government appear as two separate structures, there has been some intertwining; for monks (often of royal heritage) have commonly acted as temporal advisers, and the kings—at least in Thailand—occasionally have spent some time in the monastery. It should also be pointed out that Buddhist monastic institutions have served as a link between the rural peoples and the urban elites, helping to unify the various Theravāda countries.

Role of monastic institutions

In China the relationship between the sangha and the state has fluctuated. At times Buddhism has been seen as a foreign religion, as a potential competitor with the state, or as a drain on national resources of men and wealth. These perceptions have led to sharp purges of Buddhism and to rules curbing its influence. Some of the rules attempted to limit the number of monks and to guarantee governmental influence in ordination through state examinations and the granting of ordination certificates. At other times, such as during the early centuries of the T'ang dynasty (618–845), Buddhism was almost considered the state religion. The government created a commissioner of religion to earn merit for the state by erecting temples, monasteries, and images in honour of the Buddha.

In Japan, Buddhism has experienced similar fluctuations. During the period from the 10th to the 13th century, monasteries gained great landed wealth and temporal power. They formed large armies of monks and mercenaries that took part in wars with rival religious groups as well as in temporal struggles. By the 14th century, however, their power began to wane, and, under the Tokugawa regime that took control in the 17th century, Buddhist institutions became, to a considerable degree, instruments of state power and administration.

Only in Tibet did Buddhists establish a theocratic polity that lasted for an extended period of time. Beginning in the 12th century, Tibetan monastic groups developed relationships with the powerful Mongol khans that often gave them control of governmental affairs in Tibet. In the 17th century the Dge-lugs-pa school, working with the Mongols, established a monastic regime that was able to maintain more or less continual control until the Chinese occupation in the 1950s.

During the immediate premodern period, each of the various Buddhist communities in Asia developed some kind of working relationship with the sociopolitical system in its particular area. Within the sweep of Western colonialism and especially after the establishment of new political ideologies and political systems during the 19th and 20th centuries, these older patterns of accommodation between Buddhism and state authority were seriously challenged. In many cases bitter conflicts resulted—for example, between Buddhists and colonial regimes in Sri Lanka and Myanmar, between Buddhists and the Meiji reformers in Japan, and between Buddhists and many different communist regimes. In some cases, as in Japan, these conflicts have been resolved and new modes of accommodation have been established. In other cases, such as that of Tibet, there has been no resolution.

(Gi.T./F.E.R.)

Early Buddhist schools

THE EARLY COUNCILS

The early Buddhist councils (*sangītis*, or "recitals") were concerned largely with the purity of the faith and practice of the monastic community. Unfortunately, the councils pose an enormous problem to the historian because each major sect has its own account and opinion of them. Legend and even myth have so coloured these accounts that scholars cannot be sure when and where they took place or even who took part in them. Though many scholars deny its very existence, all Buddhist traditions maintain that a council was called at Rājagaha (modern Rājgīr) immediately after the Buddha's death. According to legend, this council (comprising 500 arhats, or accomplished monks) was responsible for the composition of the vinaya (monastic discipline), under the monk Upāli, and the *dhamma* (*i.e.,* the sutras), under the monk Ānanda, even though the latter was supposedly brought to trial at the same council. Though there were memorizers of sutras and the vinaya, as well as authorized commentators at work in the community during the period of the first three Buddhist councils, the scriptures as such existed only in an inchoate oral (yet normative) form.

Council of Rājagaha

More scholars are prone to accept the historicity of the second council that was held at Vesālī a little more than a century after the Buddha's death. According to the tradition, a controversy arose between a certain Yasa and the monks of Vajji. The 10 points of discipline observed by the Vajjian monks, and opposed by Yasa, permitted storing salt in a horn, eating in the afternoon, and drinking buttermilk after meals. These and other lax rules were condemned by the council. Many scholars believe this council to have been closely associated with the controversy that led to the open division between two segments of the early community—the Mahāsaṅghika school, which displayed more liberal attitudes, and the Sthaviravāda (Theravāda) school, which took a more conservative and elitist stance.

According to Theravāda accounts, a third council was called by King Aśoka at Pāṭaliputta (Patna) about 250 BC. Moggaliputta Tissa, president of the council, is said to have completed his *Abhidharma* treatise, the *Kathāvatthu* ("Points of Controversy"), during this council. It is also said that a controversy arose between the Sarvāstivādins and the Vibhajyavādins (usually identified with the early Theravādins) over the reality of past and future states of consciousness (*citta*s). After the Sarvāstivādin view that such states actually exist was condemned, the sect supposedly withdrew from the lower Ganges valley to Mathurā in the northwest. There it appears to have continued to develop as a transitional school between the older, more conservative schools and the nascent Mahāyāna movement.

A fourth council

According to northern Buddhist traditions, a fourth council was held under King Kaniṣka, probably in the 1st century AD, at Jalandhar or in Kashmir. This council seems to have been limited to the composition of commentaries. Because it appears that only the Sarvāstivādin viewpoint was represented, scholars generally conclude that this was a sectarian synod rather than an actual ecumenical Buddhist council. At any rate, the fourth council has never been recognized by southern Buddhists. (J.M.K./F.E.R.)

THE EIGHTEEN SCHOOLS AND THEIR SUCCESSORS

During the first several centuries of Buddhist history, at the time when the early councils were reportedly held, a number of different schools took form and developed particular traditions regarding the Buddha's teaching and its proper interpretation. In addition to the schools that scholars have connected with particular councils and the controversies associated with them, many other schools appeared. From the information available, it appears that many of these schools were loosely organized and fluid, that in some cases different names were used to designate the same group, and that in other cases the same name was applied to different groups.

According to later Buddhist tradition, 18 such schools emerged during the first few centuries of Buddhist history. (These are the schools that the practitioners of the Mahāyāna, or "Greater Vehicle," grouped together and dubbed the Hīnayāna, or "Lesser Vehicle.") Among these so-called Eighteen Schools there are a few about which some reasonably certain statements can be made. And in some cases it is possible to trace connections between the early schools and similarly oriented traditions that played a significant role in the later phases of Buddhist history.

Sthaviravāda/Theravāda. As noted above, many scholars trace the emergence of a specific "School of the Elders" (Sthaviravāda in Sanskrit, Theravāda in Pāli) to the second century following the Buddha's death. Whether or not this is true, it is clear that the Theravāda was one of the earliest of the early Buddhist schools, and it was certainly—from a historical point of view—the most successful. Because the Theravāda school ultimately became and has remained the predominant Buddhist school in Sri Lanka and much of mainland Southeast Asia, it is described in detail in the following section (see below *The major systems and their literature*).

Mahāsaṅghika/Lokottaravāda. According to traditional accounts, the Mahāsaṅghikas ("Adherents of the Great Order, or Assembly") split off from other Buddhists in the 4th or 3rd century BC, in what may have been the original schism. Their emphasis was on a more open community, a less strict version of the discipline, and a metaphysical view of the Buddha, all of which were later appropriated by the Mahāyāna (see below). In the developed form of their teaching they focused not on the historical Buddha and his teachings but on the transcendent Buddha. In their view, the Buddha is *lokottara* ("transcending the world"), indestructible, completely devoid of all worldly impurities, with stainless karma. When the Buddha utters a single word, its meaning can be understood by all creatures at a level determined by their degree of karmic purity. His body is perfect, for the body through which he reveals himself is not his true body; it is instead an apparitional body (*nirmāṇa-kāya*). Being above the world, he has boundless power and life; he neither sleeps nor dreams. Even in the state of a bodhisattva, prior to his final birth, the Buddha entered the maternal womb completely pure. All bodhisattvas can remain as long as they will among the inferior creatures for the purpose of leading creatures to salvation. The knowledge of things occurs in a single instant: all is void (*śūnya*) and without self; the ultimate end of the way of seeing is an instantaneous recognition that reveals the singular and proper character of all things.

Supramundane nature of the Buddha

The only surviving part of the Mahāsaṅghika canon, the *Mahāvastu* ("Great Subjects"), is derived from Lokottaravādins, who stem from the Mahāsaṅghikas. They assert that things of this world do not possess any reality at all. Only two principles are absolutely real, the two kinds of void (*śūnyatā*): of persons and of things. The (or a) Buddha is completely supramundane (*lokottara,* hence the name Lokottaravāda), and his historical life and actions are mere appearance, convention, or mental image.

Sarvāstivāda (P'i-t'an, Chü-she/Kusha). This group detached itself from the Sthaviravāda school, the predecessor of Theravāda, around the time of Aśoka and spread from Mathurā into Kashmir (northwestern India). Taking a different tack than the Sthaviravāda/Theravāda monks, the Sarvāstivādin scholastics developed their own set of "canonical" *Abhidharma* texts. (The *Abhidharma* [Pāli: *Abhidhamma*] texts are systematizations of the teaching in the early sutras that were composed by the monks of the various schools.) The name Sarvāstivāda refers to the doctrine that everything exists (*sarvam asti*). More specifically, this meant that all things, past and future as well as present, exist because the cognizing agent, at the moment of thinking of them, could not have contact with them if they did not exist. Karma exists, as do the component elements that constitute a human person (the five skandhas). Matter has aspects, hence characteristics, and therefore is definable. The other components have characteristics, and therefore they too are knowable. What does not exist is an atman, *pudgala,* or any other kind of underlying self or person.

A branch of this school, the Mūla-Sarvāstivāda, was widely diffused in India, Central Asia, Myanmar, Thailand, Kampuchea, and Indonesia.

The *Abhidharmakośa* of Vasubandhu

Vasubandhu's *Abhidharmakośa* ("Treasury of Abhidharma"; 4th or 5th century AD) was based on the Sarvāstivāda *Abhidharma* as interpreted in terms of its author's Sautrāntika propensities (see below *Sautrāntika/Satyasiddhi*). It became the basic text for the development of Sarvāstivāda in China and Japan, which is designated by the short form of the title: Chinese *Chü-she,* Japanese *Kusha,* for *kośa* ("treasure"). Before the Chinese translations of Vasubandhu's work in the 6th and 7th centuries, Sarvāstivāda *abhidharma* doctrine was represented in China by the P'i-t'an school, based on texts of the Gandhāra and Kashmir branch of Sarvāstivāda (*P'i-t'an* is the Chinese abbreviation for *Abhidharma*). Summarizing, interpreting, and confuting the theories of various masters in regard to external and psychosomatic *dharma*s, time, and categories, the *Abhidharmakośa* has, since its introduction, been regarded as a basic treatise on Buddhist dogma in China, as also in Tibet.

Vātsīputrīya/Sammatīya. The Vātsīputrīyas (or Pudgalavādins) probably split off sometime during the 3rd century BC. They affirmed the existence of an enduring person (*pudgala*) distinct from both the conditioned (*saṃskṛta*) and the unconditioned (*asaṃskṛta*); the sole *asaṃskṛta* for them is nirvana. The *pudgala* really exists and can transmigrate from life to life, unlike all other things, none of which possess this property. The Vātsīputrīyas refer to a text in which the Buddha speaks of a "bundle" (*i.e.,* the components of a human being, the skandhas) and of one "who carries the bundle." If consciousness exists, there must be a subject of consciousness, the *pudgala;* it is this alone that transmigrates. The Vātsīputrīyas recognized an intermediate life between death and rebirth.

The Sammatīya school, a derivation of the Vātsīputrīya school, had a wide diffusion. According to the reports of Chinese Buddhist pilgrims in the 7th century, its followers were numerous along the Ganges valley; the school flourished in Gujarāt and in eastern India, as well as in Champa, located in what is now central Vietnam; and a 16th-century Tibetan account states that it still flourished during the Pāla dynasty that ruled Bengal up to the end of the 11th century. The Sammatīyas maintained that an act vanishes but that it leaves a thrusting, a commitment to fructify, to have consequences. A person (*pudgala*) is an essence not wholly identical with its components. Thus, the Sammatīyas have tendencies toward an ontological concept of the *pudgala*—that it is real though undefinable. They also posit an intermediate existence between death and rebirth.

Sammatīya doctrines

Mahīśāsaka/Dharmaguptaka. The Mahīśāsaka school apparently took its name from its founder (but, according to others, from the name of a place). Its origins purportedly go back to a disagreement concerning disciplinary rules that occurred at the first Council of Rājagaha (483 BC); however, some of the school's theses seem to have developed later than the Sarvāstivāda.

A so-called Dharmaguptaka branch separated from the Mahīśāsaka school toward the end of the 2nd or the beginning of the 1st century BC. It recognized as its original master Maudgalyāyana (Pāli: Moggallāna), one of the first disciples of the Buddha, although its actual founder may have been Dharmagupta (3rd century BC?). It was still to be found in scattered places in the 7th century. The Dharmaguptakas added two new "baskets," or collections (*piṭakas*), a *Bodhisattvapiṭaka* and *Dhāraṇīpiṭaka,* to the regular *Tripiṭaka* of the Buddhist canon (see below). They held that the Buddha was not part of the sangha—consequently, a gift given to him is infinitely superior to one given the sangha. They emphasized the merit that derives from the cult of the stupa—a mound, often containing relics—which is a centre of devotion in lay Buddhism. They also contended that the paths of the buddha and the bodhisattva are distinct from that of the *śrāvaka* (disciple). Like the views of the closely related Mahāsaṅghika school (see above), the teachings of the Dharmaguptas contain many elements that were emphasized and developed by the Mahāyāna.

Sautrāntika/Satyasiddhi. The Sautrāntika school is so named (*sūtra* plus *anta,* "end") because it gave preeminence to the sutra portion of the canon. Its followers trace their school back to Ānanda, a close disciple of the Buddha. For them, the karmic factors (samskara) are insubstantial and momentary, disappearing as soon as they have been manifested only to reappear again to give rise to a new aggregate. There is continual motion by virtue of which a person passes from one condition to another. Every thought or act is pervaded by a very subtle impregnation that in turn is capable of impregnating the subconscious so as to generate new correlated psychic situations. The school is of great importance because its tenets were precursors of the Vijñānavāda (see below *Yogācāra/Vijñānavāda [Fa-hsiang/Hossō]*).

The Satyasiddhi school, probably derived from the earlier Sautrāntika school, is based on the *Satyasiddhi-śāstra,* a work attributed to Harivarman, a 3rd–4th century Indian writer, and known only in its Chinese version (4th–5th century). It gave birth to a school in China (Ch'eng-shih) and Japan (Jōjitsu) which maintained that all things are merely designations devoid of reality. Human beings are enveloped in the illusion that either the ego (*pudgala*) or the world (*dharmas*) is real, whereas in fact neither is. The past does not exist, the future has not yet come to be, and the present, as soon as it comes into being, disappears. Hence, the sense of continuity is illusory.

The Ch'eng-shih and Jōjitsu schools

Harivarman, like the Lokottaravādins, postulates a void, both of the *dharmas* and of the ego: no *dharmas* of any sort exist, though from the point of view of relative truth *dharmas* may appear to exist. In China this doctrine was sharply attacked by its opponents as destructive nihilism. It is perhaps improper to speak of Satyasiddhi as a school; it refers rather to certain centres that attached particular importance to the *Satyasiddhi-śāstra* without ignoring the rest of the Buddhist teachings. (In Japan, Jōjitsu is considered part of the Sanron school.)

Vinaya schools: Lü/Ritsu. The development of schools based on an emphasis on the *Abhidharma* and *Sūtra* "baskets" has been indicated above in the presentation of Sarvāstivāda and Sautrāntika. As for the *Vinaya,* however, although it constituted an integral part of the *Tripiṭaka* and regulated and consecrated the norms of monastic life, it gave rise to a school of its own only in China and Japan, apparently without any Indian precedent. The Lü-tsung (Vinaya school) originated in China in the 7th century during a period of highly intellectualist Buddhism and stressed observance of the ethical precepts and disciplinary rules. In the following century it was introduced into Japan by the famous Chinese priest Chien-chen (Japanese: Ganjin) at the invitation of the Japanese emperor and was known as Risshū or Ritsu (Japanese for Vinaya). The emphasis of Vinaya in Japan was on the correctness and validity of ordination (initiation) into the sangha, especially for monks and nuns; and controversy ensued between those who stressed the formal, external aspect and those who stressed the spiritual, internal aspect of vows and discipline. A reformed Ritsu school was established in the 13th century, based on a "self-vow discipline," marking a return to the validity of spontaneous vows, beyond any formalism, in accordance with Mahāyāna teachings. The Vinaya Lü-tsung, with its ethical, disciplinary emphasis, so congenial to the Chinese mind, continued to play a vital part in Chinese Buddhism down to modern times; Ritsu, whose teachings were in principle accepted by most Japanese Buddhists, still has its temples and following in contemporary Japan. (Gi.T./F.E.R.)

The major systems and their literature

THERAVĀDA (STHAVIRAVĀDA)

Adherents of Theravāda accept as authoritative the Pāli canon of ancient Indian Buddhism and trace their lineage back to the Sthaviras (Pāli: Theras; "Elders"), who followed in the tradition of the senior monks of the first Buddhist sangha. During the reign of the emperor Aśoka (3rd century BC), the Theravāda school traveled to Sri Lanka, where it divided into three subgroups, known after their monastic centres as the Mahāvihāravāsī, the Abhayagirivihāravāsī, and the Jetavanavihāravāsī. The Mahāvihāra form of the Theravāda tradition became dominant in Sri Lanka about the beginning of the 2nd millennium AD and gradually spread eastward, becoming established in Myanmar in the late 11th century, in Thailand in the 13th century, and in Kampuchea and Laos by the 14th century.

Beliefs, doctrines, and practices. *Cosmology.* In the Theravāda view there is a plurality of universes surrounded by water and mountain chains. Every universe has three planes: (1) the sphere of desire (*kāma-dhātu*), (2) the sphere of material form (*rūpa-dhātu*), which is associated with meditational states in which sensuous desire is reduced to a minimum, and (3) the sphere of immateriality or formlessness (*arūpa-dhātu*), which is associated with meditational states that are even more exalted and vacuous. On the plane of desire, creatures are divided into five or six species: hell beings; pretas, a species of wandering, famished ghosts; animals; human beings; gods; and a sixth group not universally acknowledged, the asuras (demigods). The matter of the world is made up of four elements: earth, water, fire, and air, held together in various combinations.

Plurality of universes, planes, and cycles

Time moves in cycles (kalpas), involving a period of involution (destruction by fire, water, air), a period of stability, a period of renewal, and a period of duration, at the end of which the destruction comes again and the cycle continues.

Human existence is a privileged state because only as a human being can a bodhisattva become a buddha. Moreover, human beings have the capability of choosing to do good works (which will result in a good birth) or bad works (which result in a bad birth); and, above all, they have the capacity to become perfected saints or even buddhas. All these capacities and activities are accounted for in terms of a series of *dhamma*s (Sanskrit: *dharmas*), instant points in continual motion or changing states, subject to appearing, aging, and disappearing.

Classification of dhammas. *Dhamma*s are divided and subdivided into many groups. The essential ones that concern the psychophysical person are the five components (skandhas; Pāli: *khandha*s), the 12 bases (*āyatana*), and the 18 sensory elements (*dhātu*). The lists converge and overlap because the teaching was codified in different ways.

The five components, or skandhas, are: (1) *rūpa,* materiality, or form, (2) *vedanā,* feelings of pleasure or pain or the absence of either one, (3) *saññā,* cognitive perception, (4) *sankhāra,* the forces that cooperate to condition the psychic activity of an individual, and (5) *viññāṇa* (Sanskrit: *vijñāna*), consciousness.

The 12 bases, or *āyatana*s, include the five sense organs and the mind (*manas*), as well as the five related sense fields and a cognizable object—that is, not an object as such but, rather, an object as it is reflected in mental perception.

The 18 elements, or *dhātu*s, comprise the five sense organs and the *mano-dhātu* (mind element), their six correlated objects, and the six consciousnesses (*viññāṇa*) of the eye, ear, nose, tongue, body, and *manas.*

Clearly in this Theravāda version, Buddhism is not concerned with metaphysical problems in the usual sense but with an analysis of the psychosomatic components of the human personality. This is because, from the Theravāda perspective, it is only through an awareness of the interrelation, combination, and operation of these components, and of the way to cultivate some and to suppress others, that a person can arrive at the state of an arhat (Pāli: *arahant;* "worthy one"). Its aim is not to promulgate a metaphysics but to liberate human beings by employing their psychic mechanism in such a way as to stop the operation of karma (Pāli: *kamma*).

The individual as a psychosomatic aggregate

Through the classification of *dhamma*s a person comes to be seen as an aggregate of many elements working together, ruled in his becoming by the law of karma, whether good or bad, and thus destined to suffer good or bad consequences. All this rests on the presupposition that there is no metaphysical entity such as an "I" or atman (Pāli: *attan*) outside of time but that there is a psychosomatic aggregate situated in time. This aggregate has the freedom of choice that allows it to perform this or that act, which can be with or without outflows, and thus capable or not capable of generating consequences.

Such classifications do not have a purely doctrinal goal; rather, they are preparatory distinctions that guide whoever accepts the teaching of the Buddha in passing from the temporal to the atemporal plane and overcoming the cycle of rebirths. Here enter the seven factors of enlightenment: clear memory, the exact investigation of the nature of things, energy and sympathy, tranquility, impartiality, and a disposition for concentration. These are assisted by subsidiaries, such as love for all living creatures, compassion, delight in that which is good or well done, and, again, impartiality. The last four are known as the "four sublime states," the necessary preconditions for liberation from karma and samsara.

The stages leading to arhatship. The Theravādins maintained that the ideal Buddhist is the arhat, the accomplished ascetic who attains nirvana through self-effort. Though the Theravādin arhat naturally "took refuge in the Buddha," his emphasis was not on the grace of the Saviour but on his *dhamma.*

In their "theology," the Theravādins tended to make distinctions. They insisted that nirvana was beyond the realm of empirical reality and that the Buddha who had founded the religion could be distiguished from the *dhamma* he taught. They maintained that monks and laymen have different roles to play both in society and in religion. The way that leads the disciple to the stages of arhatship traverses an immense number of lives, during which the aspirant gains true insight into the nature of things.

According to the Theravādins, one who gains true Buddhist insight passes through four stages. The first stage is that of the stream winner or stream enterer—*i.e.,* the one who has seen the truth, who has experienced the first real intimations of nirvana, and who will not undergo more than seven additional rebirths. The second stage is that of the once-returner—*i.e.,* the one who has moved further toward the goal so that no more than one additional rebirth will be required to attain it fully. The third stage is that of the non-returner, who will achieve complete release in the present life, or, at the very least, before another rebirth occurs. One who has reached this stage has broken free from the lower bonds: belief in a permanent self, doubt, faith in the results generated by rituals, sensual passion, and malice. The fourth and final stage is that of the arhat, who has attained complete freedom by completing all that has to be done. The arhat is free from all bonds including the desire for existence in the formed or formless worlds, as well as ignorance, excitability, and ambition.

The Buddha. The state of the Buddha, the perfectly Enlightened One, is nirvana (Pāli: *nibbāna*)—an attainment from which one does not return. It is beyond death, not caused, not born, not produced; it is beyond all becoming and devoid of all that makes up a human person. There are two kinds of nirvana. One is achieved by the Buddha while still alive, but he remains alive only until the last and most tenuous remains of karma have been expended. When these disappear, the Buddha dies and then enters the nirvana that is not burdened by any karmic residue at all.

The Buddha Metteyya

The Buddha has been given many other names, the most common of which are Arahant and Tathāgata ("He Who Has Thus Attained"). The Theravādin scriptures, in the later stages, express a belief in previous buddhas before Gotama (six in one list, more in others) and also in a future buddha, Metteyya (Sanskrit: Maitreya), who presently dwells in the Tuṣita Heaven and who will come into the world when the proper time arrives.

Meditation. In the Theravāda tradition two basic forms of meditation (Pāli: *jhāna;* Sanskrit: *dhyāna*) have been practiced in various forms and combinations. The first of these is closely related to a Hindu tradition of yoga practice involving a process of moral and intellectual purification associated with four stages of *jhānic* attainment. In the Theravāda context the meditator achieves detachment from sensual desires and impure states of mind through analysis and reflection and thereby attains an emotional state of satisfaction and joy. In the second stage, intellectual activities are abated to a complete inner serenity; the mind is in a state of "one-pointedness" or concentration, joy, and pleasantness. In the third stage, every emotion, including joy, has disappeared, leaving the meditator indifferent to everything while remaining completely conscious. The fourth stage is the abandoning of any sense of satisfaction, pain, or serenity because any inclination to a good or bad state of mind has disappeared. The meditator thus enters a state of supreme purity, indifference to everything, and pure consciousness.

At this point the meditator begins the *samāpatti*s (or the higher *jhānic* attainments). Beyond all perception of form, withdrawn from the influence of perception, immune to the perception of plurality, concentrating on infinite space, the meditator reposes in the condition of spacial infinity. Going beyond this stage, the meditator concentrates on the limitlessness of consciousness and attains it. Proceeding further and concentrating on the nonexistence of everything whatsoever, he achieves a state in which there is absolutely nothing. Even further on, the meditator attains the highest level of realization in which there is neither perception nor nonperception.

The second form of Theravāda meditation is called *vipassanā,* or insight meditation. This kind of meditation requires concentration (produced by exercises such as concentrating on one's breathing), which lead to one-pointedness of mind. This one-pointedness of mind is then used to attain—directly—Buddhist insight into the saving truth that all reality is without self and impermanent and is filled with suffering, even the exalted *jhānic* states of consciousness. This insight, from the Buddhist perspective, gives direct access to progress along the path and to the actual attainment of nirvana itself.

In the classical Theravāda texts the emphasis is placed on the *jhānic* forms of meditation, though the *vipassanā* forms are never completely ignored. In recent years, however, there has been an increasing emphasis on practices in which the *vipassanā* approach is predominant.

(Gi.T./F.E.R.)

The Three Baskets

The Pāli canon (*Tipiṭaka*). The earliest systematic and most complete collection of early Buddhist sacred literature is the Pāli *Tipiṭaka* ("The Three Baskets"; Sanskrit: *Tripiṭaka*). Its arrangement reflects the importance that the early followers attached to the regulations of the monastic life (*Vinaya*), to the discourses of the Buddha (*Sutta*), and subsequently to the interest in scholasticism (*Abhidhamma*).

By courtesy of the Newberry Library, Chicago

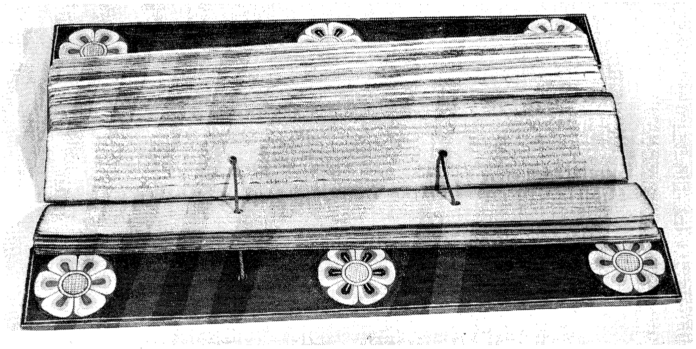


Figure 2: Buddhist Pāli manuscript from Kandy, Sri Lanka, about 45 centimetres (18 inches) long. The palm-leaf pages are threaded with twine, and the covers are wood with painted decoration. In the Newberry Library, Chicago.

The Pāli *Vinaya Piṭaka,* which is still in theory the rule in Theravāda monasteries, although large sections have fallen into disuse, is divided into five major parts grouped into three divisions—*Sutta-vibhaṅga* ("Division of Rules"), *Khandhaka*s ("Sections"), and *Parivāra* ("Accessory").

The *Sutta-vibhaṅga* is a commentary on the *Pātimokkha-sutta* ("Obligatory Rules"), which forms the nucleus of the *Vinaya Piṭaka.* It is one of the oldest parts of the Pāli canon and utilizes an archaic language. It consists of two parts, the *Bhikkhu-pātimokkha* ("Rules for Monks") and the *Bhikkhunī-pātimokkha* ("Rules for Nuns"). The offenses a monk or nun is likely to commit are listed according to their gravity, as, for instance, those warranting permanent expulsion from the order (*pārājika*), those entailing temporary suspension (*sanghādisesa*), and those of which a guilty person can be absolved by a formal confession before the monastic order (*pācittiya*). The commentary on the *Pātimokkha* is divided into the *Mahā-vibhaṅga* ("Great Division"; 227 rules for monks) and *Bhikkhunī-vibhaṅga* ("Division Concerning Nuns"). The latter lists additional specific rules and regulations.

The *Khandhaka* section of the *Vinaya* consists of two parts, the *Mahāvagga* ("Great Grouping") and the *Cullavagga* ("Small Grouping"). The topics dealt with in these two sections have not always been clearly distinguished and also lack logical sequence. They contain rules for ordination; rules for "observance days," when all monks resident within the district were required to assemble for recitation of the *Pātimokkha;* descriptions of rainy-season retreats, clothing, food, and medicines; judicial rules; rules for the instruction of nuns; and so forth. While the *Mahāvagga* is presented as a kind of history of the developing Buddhist community, the *Cullavagga* supplements the details of the former to make an authoritative compilation of the Buddha's sayings regarding the discipline.

The *Parivāra* contains summaries and classifications of the disciplinary rules. It is a later supplement intended not only to help monks and nuns to remember the rules but also to make them aware of the circumstances that would bring them within the orbit of these rules.

Sutta Piṭaka

By far the largest of the three "baskets" is the *Sutta Piṭaka* ("Basket of Discourse"), which consists of five collections (*nikāya*s) containing the discourses attributed to the Buddha. Although, from a literary viewpoint, many of the discourses seem to be drawn out and repetitive, they nevertheless make rewarding reading because of the sublimity of thought and the richness and beauty of the illustrative similes that they contain. The discourses, reported by the Buddha's disciples, begin with the affirmative statement "Thus I have heard" and then relate the place and occasion of the discourse. At the end they affirm that the listeners are delighted and that they rejoice in what the Buddha has said. It is obvious that these discourses do not represent the exact words of the Buddha, although some phrases may have been accurately remembered. Still, they reveal the personality, the didactic technique, and the spirit of the founder. The discourses are chiefly in prose, except for stanzas illustrating or summing up a particular point.

The grouping of the discourses into *nikāya*s does not rest on any kind of topical basis. Apparently there existed two groups of teachers (*bhāṇaka*s), who memorized certain *sutta*s ("discourses," or "sermons") and handed them down to their disciples orally. Reciters of lengthy verses were called Dīghabhāṇakas, and reciters of middle-length verses Majjhimabhāṇakas. The third and fourth *nikāya*s (*Saṃyutta* and *Aṅguttara*) seem to reflect a later development, their aim being to rearrange the topics dealt with in the *Dīgha* and *Majjhima Nikāya*s.

The *Dīgha Nikāya* ("Collection of Long Discourses") contains 34 *sutta*s, some of considerable length, presenting a vivid picture of the different aspects of life and thought at the Buddha's time. Divided into three books, it contrasts superstitious beliefs, various doctrinal and philosophical speculations, and ascetic practices with Buddhist ethical ideas, which are elucidated with the help of similes and examples taken from the everyday life of the people. One of the most interesting *suttanta*s ("discourses") is the *Mahāparinibbāna Sutta,* which gives an account of the last days of the Buddha and stresses the importance of striving for emancipation.

The *Majjhima Nikāya* ("Collection of the Middle Length Sayings") contains 152 *sutta*s in its present version, while the Chinese one, preserving the lost Sarvāstivāda collection, has 222, some of which are also found in other *nikāya*s of the Pāli canon. Like the *Dīgha,* the *sutta*s in the *Majjhima* present Buddhist ideas and ideals, illustrating them by similes of great literary beauty.

The *Saṃyutta Nikāya* ("Collection of Kindred Discourses") has altogether 2,941 *sutta*s, classed in 59 divisions (called *saṃyutta*) grouped in five parts (*vagga*). The first *vagga* has *sutta*s that contain stanzas. The *sutta*s begin with a description of the particular occasion when the stanzas were spoken; the stanzas themselves represent a kind of questioning and answering. The second *vagga* deals with the important principle of dependent origination—the chain of cause and effect affecting all things. The third *vagga* presents the anatman (no-self) doctrine, which is the rejection of an abiding principle that could be termed a self or a pure ego. The fourth *vagga* is very similar to the previous one, but here it is not the philosophical principle underlying the analysis that is stressed but the transitoriness of the elements constituting reality. The fifth *vagga* is devoted to a discussion of the basic principles of Buddhist philosophy, religion, and culture.

The *Aṅguttara Nikāya*

The *Aṅguttara Nikāya* ("Collection of the Gradual Sayings") contains as many as 2,308 small *sutta*s arranged according to the number of topics discussed, ranging from one to eleven. One *sutta* relates that loving kindness practiced for a fraction of a second only will yield great merit. Other *sutta*s state that there are three areas in which training is needed—in conduct, concentration, and insight—and that there are eight worldly concerns—gain, loss, fame, blame, rebuke, praise, pleasure, and pain. Here, too, similes enliven an otherwise dry presentation.

The *Khuddaka Nikāya* ("Collection of Small Texts") comprises 15 separate titles:

1. *Khuddaka-pāṭha* ("Small Reading"). This is the smallest book in the entire *Tipiṭaka.* Compiled for use by primary trainees, its contents are used on various occasions. Two *sutta*s have been borrowed from the *Suttanipāta* (see

below), and their recitation is regarded as very auspicious.

2. *Dhammapada* ("Verses on the *Dhamma*"). This work contains 423 verses in 26 chapters. Presenting the maxims of Buddhist ethics, it not only occupies an eminent place in the religious life of the peoples in Buddhist countries but is also of universal appeal, as it recommends a life of peace and nonviolence and declares that enmity can never be overcome by enmity, only by kindness.

3. *Udāna* ("Utterances"). This contains 80 utterances attributed to the Buddha or his chief disciples, when they had achieved the bliss of their emancipation or spoke in appreciation of a sublime state.

4. *Itivuttaka* ("Thus Said"). This contains 112 short pieces dealing with ethical principles, such as generosity, good and evil, greed, passion, and malice.

5. *Suttanipāta* ("Collection of *Sutta*s"). This is one of the oldest Buddhist texts in existence today. It is partly in verse, partly in a mixed style of prose and verse. The verse part is of high poetic quality.

6. *Vimānavatthu* ("Tales of Heavenly Mansions"). This book describes the different abodes of deities, male and female, who are born in the heavens as a result of their former meritorious deeds.

7. *Petavatthu* ("Tales of Ghosts"). This work gives an account of the various purgatories and the woes of the beings reborn there as a result of their wicked deeds.

8. *Theragāthā* ("Hymns of the Elders"). This collection contains songs attributed to 264 personal disciples of the Buddha. The songs are said to have been composed when their authors experienced the bliss of emancipation.

9. *Therīgāthā* ("Hymns of the Senior Nuns"). These are the songs attributed to about 100 female disciples of the Buddha. They provide rich material for the study of the position of women at the time of the Buddha. Their merit consists in their revealing the deep impression the Buddha's teaching made upon their life. A personal tone is unmistakable.

10. *Jātaka* ("Lives [of the Buddha]"). Only the verses are considered to be canonical, while the 547 tales of the Buddha's previous lives are considered a later addition. The prose stories contain legends, fables, humorous anecdotes, and short sayings, as well as lengthy romances.

11. *Niddesa* ("Exposition"). This work, consisting of two parts, *Mahāniddesa* and *Cullaniddesa,* actually belongs to the group of commentaries. The last two chapters comment on the *Suttanipāta.*

12. *Paṭisambhidā-magga* ("The Way of Analysis"). This is a kind of encyclopaedia of the philosophical ideas in the *Sutta Piṭaka.* It is primarily meant for reference and intensive study.

13. *Apadāna* ("Stories"). This is a collection of stories of the previous lives of the Buddha, the pratyeka buddhas (who attain enlightenment by themselves and are unconcerned about the enlightenment of others), and the arhats of the early Buddhist sangha, whose *Theragāthā* and *Therīgāthā* songs are incorporated and embellished with rich biographical detail. The concluding sentence of each *apadāna* in the collection is intended to show that even the smallest meritorious act has the potentiality of giving vast positive results even after a long time. All the stories are in verse.

14. *Buddhavaṃsa* ("Lineage of the Buddha"). This work relates the lives of 24 previous buddhas, of Gotama (the historical Buddha), and of Metteyya (Sanskrit: Maitreya; the future buddha). According to the text, the stories are told by the historical Buddha himself.

15. *Cariyā Piṭaka* ("Basket of Conduct"). This collection retells 35 Jātakas (stories of the Buddha's previous lives) in verse form, illustrating the bodhisattva's practice of the 10 perfections (paramitas) necessary for the attainment of Buddhahood.

Abhidhamma Piṭaka

The third of the three "baskets," the *Abhidhamma Piṭaka* ("Basket of Scholasticism"), comprises seven works that, although based on the contents of the Buddha's discourses, deal with selected and specific topics which form the basis for the later philosophical interpretations. The Pāli version is a strictly Theravāda collection and has little in common with the *Abhidhamma* works recognized by other schools of Buddhism. It consists of seven works: (1) *Dhammasaṅgaṇi* ("Summary of *Dhamma*"), an enumeration of the entities constituting reality, (2) *Vibhaṅga* ("Division"), a definition of these entities from various points of view, (3) *Dhātukathā* ("Discussion of Elements"), a classification of the elements of reality according to various levels of organization, (4) *Puggalapaññatti* ("Designation of Person"), an interesting psychological typology in which people are classified according to their intellectual acumen and spiritual attainments, (5) *Kathāvatthu* ("Points of Controversy"), a later work discussing the controversial doctrinal points among the various ancient schools, (6) *Yamaka* ("Pairs"), dealing with basic sets of categories arranged in pairs of questions, and (7) *Paṭṭhāna* ("Activations"), a voluminous work discussing 24 kinds of causal relations.

Early noncanonical texts in Pāli. The noncanonical literature of Theravāda Buddhism consists, to a large extent, of commentaries on the *Tipiṭaka* texts but also includes independent works. Among the Pāli writers and exponents of ancient Buddhism who attempted to harmonize the apparently conflicting teachings and to grasp the inner meaning of the doctrine, four names stand out—Nāgasena, Buddhadatta, Buddhaghosa, and Dhammapāla. Nothing is known of Nāgasena except that he was the learned monk who debated with the Greco-Bactrian ruler Menander, as described in the famous literary work *Milinda-pañha.* Buddhadatta and Buddhaghosa were 5th-century contemporaries, deeply versed in the Pāli tradition, while Dhammapāla was slightly later but followed the same tradition.

Nāgasena is supposed to have compiled the *Milinda-pañha* ("Questions of King Menander"). It is generally assumed that this work was written either at the time of Menander (*c.* 150 BC) or shortly afterward, but certainly before the time of Buddhaghosa (AD 400), who very often quotes this work as an authority.

The *Milinda-pañha* is one of the great literary achievements in the field of Indian prose writing. The author begins his work with an account of the past lives of himself and King Menander; it is because of events in these past lives that the two are to meet again in this life. Menander, a well-informed scholar and keen debater, was disheartened when he could find no one to solve his problems regarding Buddhist teachings. But one day he saw Nāgasena going on his begging round. The monk's serenity made a deep impression on the king, who visited him in his monastery. They had a conversation that was later resumed at the palace and that forms the subject matter of the *Milinda-pañha.* The *Milinda-pañha* presents a profound and comprehensive exposition of Buddhist doctrine, ethics, and psychology. This work, like several others, contains the famous statement that, just as the parts of a chariot put together in a specific way constitute the chariot and there is no chariot as such over and above its parts, similarly the different components of an individual make up the individual and there is no other additional entity to hold the components together.

Writings of Buddhaghosa

Buddhaghosa (fl. early 5th century AD) is undoubtedly the most prolific and important writer in the Pāli language. There are different views about his birthplace. Certainly, he stayed at Bodh Gayā, in eastern India, for a long time, and this brought him into contact with Sinhalese monks, because the vihara (monastery) there had been built with the permission of Emperor Samudra Gupta (*c.* AD 330–380) for Sinhalese pilgrims. Buddhaghosa moved from Bodh Gayā to Sri Lanka and stayed at the Mahāvihāra ("Great Monastery") in Anurādhapura,which possessed a rich collection of commentarial literature, probably in Old Sinhalese. The first work that Buddhaghosa wrote was the *Visuddhimagga* ("Way to Purity"), a compendium of Theravāda teaching that has been greatly revered by his successors. No chronological sequence can be established for his other works. Using the *Mahā-aṭṭhakathā* ("Great Commentary"), he wrote commentaries on the *Vinaya,* the first four *nikāya*s, and the seven books of the *Abhidhamma Piṭaka.*

A number of other works are traditionally attributed to Buddhaghosa, although modern scholarship indicates that he was not the author. These works include commentaries on the *Suttanipāta* and the *Khuddaka-pāṭha,* as well as the extremely important commentaries on the *Dhammapada* and the *Jātakas*. The commentary on the *Jātakas* has as its introduction what is perhaps the most famous "biography" of the Buddha in Pāli—a biography that begins with the hero's vow, made in a previous life, to become a buddha, and concludes with his coming to the Jetavana monastery where he purportedly was staying when he recounted the 547 stories that follow. These stories, which all recount events in one of the Buddha's previous lives, range from very brief narratives to full-scale romances (for example, the extremely famous *Mahāvessantara Jātaka,* which recounts the story of the Buddha's last life before his birth as Siddhattha, during which he perfected the virtue of sacrificial giving). In all Theravāda countries these narratives and romances have exerted a tremendous influence on everything from the fine arts to law.

Buddhadatta, a contemporary of Buddhaghosa, was a native of Uragapura, near modern Tiruchirāppalli, Tamil Nadu, southern India. Like Buddhaghosa, he went to Sri Lanka to study at the Mahāvihāra in Anurādhapura. On his return he wrote his works in a monastery on the banks of the Kaveri River. His *Abhidhammāvatāra* ("The Coming of the *Abhidhamma*"), although a summary of the older commentaries on the *Abhidhamma Piṭaka,* stands supreme. In no way did he follow Buddhaghosa blindly. He reduced Buddhaghosa's five metaphysical ultimates—*i.e.,* form, feeling, sensations, motivations, and perception—to four, namely, mind, mental events, forms, and nirvana. This creative classification, similar to that of the Sarvāstivādins, makes him a philosopher in his own right rather than a commentator who merely restates things in new terms.

Dhammapāla, who probably came from southern India, is credited with having written commentaries on all the works left untreated by Buddhaghosa, whom he mentions in his work the *Paramattha dīpanī* ("Elucidation of the True Meaning"), a commentary on several books of the *Khuddaka Nikāya.* In the *Paramattha Mañjūsā* ("Jewel Box of the True Meaning"), his commentary on Buddhaghosa's *Visuddhimagga,* he quotes a verse from the Hindu scripture *Bhagavadgītā* and frequently mentions the views of other schools and teachers. This work provides valuable information about intellectual activity in traditional circles.

Sinhalese chronicles

At the close of the 4th century AD there existed in Sri Lanka an older work, a kind of chronicle, of the history of the island from its legendary beginning onward. It probably formed part of the *Mahā-aṭṭhakathā,* the commentarial literature that formed the basis of the works by Buddhaghosa and others. The accounts contained therein are reflected in the *Dīpavaṃsa* ("History of the Island"), which appears to be a poor redaction in Pāli of an older Old Sinhalese version. The *Mahāvaṃsa* ("Great Chronicle") by Mahānāma, continued in the *Cūlavaṃsa* ("Little Chronicle"), shows much greater skill in the use of the Pāli language and makes liberal use of other material. It is an artistic composition containing rich mythic, legendary, and historical material. The *vaṃsa* tradition continued in Sri Lanka and other Theravāda countries and is still alive in Sri Lanka.

Later Theravāda literature. During and after the "revival" and spread of the Theravāda in the early centuries of the 2nd millennium AD, a new corpus of Theravāda literature came into being. This literature includes commentaries and independent works written in Pāli in Sri Lanka and the Theravāda countries of Southeast Asia (for example, the highly respected commentary on the *Maṅgala Sutta* written in northern Thailand in the 16th century). It also includes many important texts written in vernacular languages, including Sinhalese, Burmese, Thai, Laotian, and Khmer. One classic example is the 14th-century cosmology called the *Traiphūmikathā* (*Three Worlds According to King Ruang*), which is the oldest known full-length text written in Thai. (H.G./F.E.R.)

MAHĀYĀNA

Arising in India, the Mahāyāna version of Buddhism spread to Central Asia, China, Japan, mainland Southeast Asia, Java, Sumatra, and even Sri Lanka (Abhayagiri monastery). It became the Pan-Asiatic form of Buddhism and involved basic shifts in doctrine and approach for which, however, there were precedents in earlier schools. It taught that neither the self nor the *dharmas* exist. Moreover, for the elite arhat ideal, it substituted the bodhisattva—*i.e.,* the one who possesses the innate tendency to become a buddha, a disposition inherent in all persons. In Mahāyāna, love for creatures is exalted to the highest; a bodhisattva is encouraged to offer the merit he derives from good deeds for the good of others. The tension between morality and mysticism that agitated India also entered the Mahāyāna.

Nature and characteristics. Mahāyāna is not merely a metaphysics, dealing with the basic structure and principles of reality. It is also and primarily a theoretical propaedeutic to the achievement of a desired state or condition. Thus there is a coexistence of theoretical investigation and supreme experience: the former, the premise; the latter, the consequence. The convergence of meditative exercises leads to an emptying of thought to reach a point in which one proceeds from voidness to voidness and finally to the ultimate where even the most attenuated thought vanishes. Rational activity is exercised until it becomes quiescent: prajna itself, the supreme wisdom, by successive emptying becomes nullified, and only in doing so does it identify with the unutterable ultimate reality.

The doctrine of voidness

Basic teachings. *The Buddha: divinization and multiplicity.* In the Mahāyāna tradition, the Buddha is viewed not merely as a human master and model but also as a supramundane being. He multiplies himself and is reflected in a pentad of buddhas: Vairocana, Akṣobhya, Ratnasambhava, Amitābha, and Amoghasiddhi. Some of these, taking the place of Śākyamuni, are revealers of doctrines and elaborate, complicated liturgies.

As Mahāyāna developed, a great deal of literature called *Buddhavacana* (Revelation of the Buddha) was circulated, but it went far beyond the ancient canons; it was proposed as the highest revelation, superseding prior texts. In this literature the teaching is viewed not as merely of one kind but as on various levels, each adapted to the intellectual capacity and karmic propensities of those who hear it. The Buddha is no longer simply the historical sage of the Śākyas but is now supramundane (*lokottara*). Even the sangha is of two types: that of this world and that beyond it.

The bodhisattva ideal. The essential premise of the bodhisattva ideal is to generate in one's own self the thought of enlightenment and to fulfill the vow to become a buddha, foregoing entrance into nirvana in order to remain in the world as long as there are creatures to be saved from suffering. With that vow the aspirant begins the career of a bodhisattva, which traverses 10 stages or spiritual levels (*bhūmi*) and achieves purification through the practice of the 10 perfections (paramitas). These levels, which become progressively higher, elevate the bodhisattva to the condition of a buddha. The first six levels are preliminary, representing the true practice of the six perfections (generosity, morality, patience, vigour, concentration, and wisdom). Irreversibility occurs as soon as the seventh stage is reached. From this moment the bodhisattva assumes the true buddha nature, even though further purification and fortification must be achieved in the stages that follow. This is the moment when, having performed his duty, he engages in activity aimed at completely fulfilling the obligations of a bodhisattva. The difference between this and the preceding six stages is that now the activity is explained as an innate and spontaneous impulse manifested unconstrainedly and therefore not subjected to doubts. Everything is now uncreated, ungenerated; thus, the body of the bodhisattva becomes identified more and more completely with the essential body (*dharma-kāya*), with buddhahood, and with omniscience.

The three Buddha bodies. The three bodies (*tri-kāya; i.e.,* modes of being) of the Buddha, which became a subject of major discussion in the Mahāyāna, are rooted in the Theravāda teachings concerning the physical body (which consists of four elements), the mental body, and the body of the law. It is with the Mahāyāna, however, that the theory of the three bodies enters into the salvation

The phenomenal body

process and assumes central significance in the doctrine. The phenomenal body (*nirmāṇa-kāya*) is a manifestation of the Buddha among creatures to teach them the path to liberation—a body that for some schools is nothing but an illusory appearance of eternal reality. The enjoyment (or bliss) body (*sambhoga-kāya*) is the body to which contemplation can ascend. At the higher stages of supramundane contemplation that body manifests to the bodhisattva its splendour and reveals doctrines unintelligible to those who are unenlightened. The unmanifested body of the law (*dharma-kāya*) already appears in the *Saddharmapuṇḍarīka,* or *Lotus Sutra,* a transitional text that became central in many Mahāyāna devotional schools (see below *Saddharmapuṇḍarīka* and *Nichiren*). In many Mahāyāna texts buddhas are infinite, and all partake of an identical nature—the *dharma-kāya.*

As anticipated in ancient schools, the Buddha is the law (*dharma*). "He who sees the law sees me; he who sees me sees the law." There is identification of the Buddha with an eternal *dharma,* with enlightenment (bodhi), and hence with nirvana; later, real existence will be opposed to the mere appearance of existence, and voidness, the "thingness of things," an undefinable condition, present and immutable within the Buddhas, will be stressed. All is in the *dharma-kāya;* nothing is outside of it, just as nothing is outside of space; transcendence and immanence come together. Other schools posit a presence that is innate within all human beings, even if it is not perceived. It is like a gem hidden in dross, which shines in its purity as soon as the veil of ignorance is removed.

New revelations. New revelations are made not only to human beings on earth but also in the heavenly paradises by Śākyamuni and other buddhas. The teaching is expounded uninterruptedly in the universe because worlds and paradises are infinite and all buddhas are consubstantiated with the essential body. The assemblies to which they speak consist not only of *śrāvaka*s (disciples) but also of bodhisattvas, gods, and demons. The authors of the new doctrines were captivated by exaltations that often make their discourses logically implausible: phantasmagoria of celestial choruses, fabulous visions in which shine flashes of new speculations, and trains of thought under the influence, more or less conscious, of speculative and mystical Indian traditions. The texts, from which new trends spring, overflow with repetitions and modulate the same arguments with a variety of readings.

The task of Mahāyāna thinkers was very difficult because it was not easy to produce a completely logical arrangement from this prolix literature. The appearance of some of these books is surrounded with legend. The *Prajñāpāramitā* and the *Avataṃsaka-sūtra*s, for instance, are said to have been concealed by the nagas, demigods living at the bottom of lakes and rivers, in miraculous palaces. There are various *Prajñāpāramitā* ("Perfection of Wisdom") texts, ranging from 100,000 verses (the *Śatasāhasrikā*) to only a few lines (the *Prajñāpāramitāhṛdaya-sūtra,* famous in English as the *Heart Sutra*). The *Prajñāpāramitā-sūtra*s announce that the world as it appears to us does not exist, that reality is the indefinable "thingness of things" (tathata; *dharmāṇām dharmatā*), that voidness (*śūnyatā*) is an absolute "without signs or characteristics" (*animitta*).

Relative and absolute truth

The fundamental assumption of the *Prajñāpāramitā* is expounded in a famous verse: "like light, a mirage, a lamp, an illusion, a drop of water, a dream, a lightning flash; thus must all compounded things be considered." Not only is there no "self," but all things lack a real nature (*svabhāva*) of their own. There are two truths: relative truth, which "applies to things as they appear," and absolute truth, the intuition of voidness (it can be of 10, 14, 18, or 20 kinds).

The Mahāyāna schools and their texts. Mahāyāna comprises the following main schools: the Mādhyamika; the Yogācāra or Vijñānavāda (Vijñaptamātratā); the Avataṃsaka; the school of the identity of the paths to salvation (*ekāyana*) represented by the *Saddharmapuṇḍarīka* ("Lotus of the True Law"; the *Lotus Sutra*); the various devotional (Pure Land) schools; and the Dhyāna school (Ch'an in China, Zen in Japan).

Mādhyamika (San-lun/Sanron). The Mādhyamika ("Doctrine of the Middle Way"), also known as Śūnyavāda ("Theory of Negativity or Relativity"), system—which held both subject and object to be unreal—is the systematized form of the doctrine of *śūnyatā* (cosmic emptiness) contained in the *Prajñāpāramitā* literature. The most famous exponent of this system was the Indian philosopher Nāgārjuna (*c.* AD 150–*c.* 250), the presumed author of the voluminous *Mahāprajñāpāramitā-śāstra* ("The Great Treatise on the Perfection of Wisdom"), preserved in its Chinese translation (402–405) by Kumārajīva; of the *Mūlamadhyamakakārikā* (more commonly known as *Mādhyamika Kārikā;* "Fundamentals of the Middle Way"), which is the Mādhyamika work par excellence; of the *Śūnyatāsaptati,* expounding the unreality of all elements of reality; of the *Vigrahavyāvartanī,* a refutation of possible objections to the doctrine of *śūnyatā;* of the *Vyavahārasiddhi,* teaching that absoluteness and relativity can coexist in practice; and of the *Yukti-ṣaṣṭika,* dealing with relativity. It is possible that Nāgārjuna is also the author of the *Ratnāvalī,* of the *Pratītyasamutpādahṛdaya,* and of the *Sūtrasamuccaya,* besides many other works attributed to him. Nāgārjuna's chief pupil was Āryadeva. His main work, the *Catuḥśataka,* criticizes both other forms of Buddhism and the classical Sanskrit philosophical systems. Together with Nāgārjuna, Āryadeva is the real founder of the Mādhyamika system.

Nāgārjuna and his followers attempted to arrive at a middle position, devoid of name and character and beyond all thought and words. They began by employing a rigorous logic to demonstrate the absurdity of various philosophical positions, including those of the Hindus and those of other Buddhists. On the assumption that any contradiction is proof of error, Nāgārjuna took any point of view that would reveal the error of his opponents. Yet, he did not therefore accept the opposing point of view but only used it as a means to show the relativity of the system he was attacking. He was just as willing to refute his first position. In this way he claimed adherence to no doctrine.

The middle path

With this method of reduction to absurdity or to contradiction, Nāgārjuna attempted to prove that all worldly thought is empty (*śūnya*) or relative, and to point to his belief that the true path is that of the middle, the path that is between or, more correctly, above extremes. This belief has been called the doctrine of emptiness of all things; however, as has been pointed out, this too is relative and should be seen only as a means of argumentation, which must itself be transcended.

Nāgārjuna presented this middle path above extremes most clearly in the following statement of what he considered to be the Eightfold Truth of Buddhism:

> Nothing comes into being, nor does anything disappear. Nothing is eternal, nor has anything an end. Nothing is identical, nor is anything differentiated. Nothing moves here, nor does anything move there.

In presenting these pairs of opposites, Nāgārjuna taught that anything that can be conceptualized or put into words is relative. This led to the Mādhyamika identification of nirvana and samsara. Both are empty concepts with the truth lying somewhere beyond.

After the emptiness or relativity of the world has been proved, the question arises as to how one is to go beyond this position. Nāgārjuna answered that humans are not irreconcilably caught in this world, for this world can be used as a ladder leading to the absolute—beyond all duality. The transition that can be effected from this world to salvation has been called Nāgārjuna's doctrine of two truths. The relative truth is of this existence. By the logical method, all propositions can be destroyed. This leads to the realization that all is emptiness and from this to the intuition of an absolute truth beyond all conceptions. The link between these two truths—the relative and the absolute—is the Buddha. He experienced the absolute truth, which is *niṣprapañca—i.e.,* inexplicable in speech and unrealizable in ordinary thought—and yet he returned to point to this truth in the phenomenal world. By following this path, one can be saved.

Thus, Nāgārjuna taught that through the middle path of Mādhyamika, which is identified as the Buddha's true teachings, one is guided to an experience beyond affirmation and negation, being and nonbeing. Mādhyamika is a philosophy that can rightly be called a doctrine of salvation, for it claims to present humans with a system that leads to rescue from their situation.

A new phase in the development of the Mādhyamika system was initiated during the 5th and 6th centuries when two subtraditions became differentiated. The more conservative Prāsaṅgika school, which emphasized a more negative form of argumentation, was founded by Buddhapālita (*c.* 470–540), who wrote, among many other works, a commentary on Nāgārjuna's *Mādhyamika Kārikā.* The school was continued by Candrakīrti, a famous logician of the 7th century, who wrote another commentary on the same text; and by Śāntideva (7th–8th centuries), who wrote two of the most popular works in all Mahāyāna literature: the *Śikṣā-samuccaya* ("Summary of Training") and the *Bodhicaryāvatāra* ("The Coming of the Bodhisattva Way of Life"). The more liberal Svātantrika school, which utilized a more syllogistic mode of argument, was founded by Bhāvaviveka, a contemporary of Buddhapālita, who also wrote a commentary on Nāgārjuna's *Mādhyamika Kārikā.* The school was continued by Śāntirakṣita, a great scholar who wrote the *Tattvasaṃgraha* ("Summary of Essentials") and the *Mādhyamikālaṅkāra Kārikā* ("Verses on the Ornament of the Mādhyamika Teaching"), on which his disciple wrote a commentary. Both the Svātantrika tradition and the Prāsaṅgika tradition (reasserted by Atīśa) had a great influence on Buddhist philosophy in Tibet.

Introduction into China

The Mādhyamika school of thought was spread to China from India by Kumārajīva, a missionary translator of Indian-Kuchan parentage, in the 5th century. Three of the texts that he translated from Sanskrit into Chinese—the *Mādhyamika Kārikā* and the *Dvādaśamukha-śāstra* or *Dvādaśa-dvāra-śāstra* ("The Twelve Topics or Gates Treatise") of Nāgārjuna and the *Śata-śāstra* ("One Hundred Verses Treatise") of Āryadeva—became the basic texts of the Chinese San-lun (Japanese Sanron) or "Three Treatise" school of Mādhyamika. For a brief period this school was challenged by a more positivistic form of Mādhyamika called the Ssu-lun, or "Four Treatise," school, which also accepted the *Mahāprajñāpāramitā-śāstra* as a basic text. This school, however, was soon overwhelmed by San-lun as taught by Kumārajīva's disciple, Seng-chao, and later by Chi-tsang. Both of these Chinese Mādhyamika masters restated Nāgārjuna's thesis in numerous influential commentaries. A Korean disciple of Chi-tsang named Ekwan (Hui-kuan) then spread San-lun (Korean Samnŏn) to Japan in 625, thus completing the rapid spread of Mādhyamika thought from India to China and to Japan. This school, despite its profound and widespread influence, never gained popularity among the masses; it remained rather the basis for logical and philosophical thought among the learned few, rarely forming a separate or independent sect.

Yogācāra/Vijñānavāda (Fa-hsiang/Hossō). The Yogācāra (or Vijñānavāda) school is traditionally ascribed to the brothers Asaṅga and Vasubandhu (5th century AD), to whom may be added Sthiramati (6th century). These writers were systematizers of doctrines already being taught and contained in such literature as the *Laṅkāvatāra-sūtra* and the *Mahāyāna-śraddhotpada-śāstra* (attributed to Aśvaghoṣa but probably written in Central Asia or in China). Yogācāra explored and propounded basic doctrines that were to be fundamental in the future development of Mahāyāna and that influenced the rise of Tantric Buddhism (see below).

Its central doctrine is that only consciousness (*vijñāna-mātra;* hence the name Vijñānavāda) is real, that thought or mind is the ultimate reality. External things do not exist; nothing exists outside the mind. The common view that external things exist is due to an error or misconception that is removable through a meditative or yogic process that brings a complete withdrawal or "revulsion" from these fictitious externals and an inner concentration and tranquility.

Store consciousness (*ālaya-vijñāna*)

A store consciousness (*ālaya-vijñāna*) is postulated as the receptacle, or storehouse, of the imprint of thoughts and deeds, the *vāsanā* (literally, "dwelling") of various karmic seeds (*bīja*s). The "seeds" develop into touch, mental activity, feeling, perception, and will, corresponding to the five skandhas. Then ideation (*manas*) develops, which sets off a self or mind against an outer world. Finally comes the awareness of the objects of thought via sense perceptions and ideas. The store consciousness must be purified of its subject-object duality and notions of false existence and restored to its pure state. This pure state is equivalent to the absolute "suchness" (tathata), to buddhahood, to the undifferentiated.

Corresponding to these three elements of false imagination (*vikalpa*), right knowledge, and suchness are the three modes in which things are: (1) the mere fictions of false imagination, (2) the relative existence of things, under certain conditions or aspects, and (3) the perfect mode of being, corresponding to right knowledge. The latter state of consciousness and being is that ultimately attained by the bodhisattva in buddhahood. Corresponding to this threefold version of the modes of being and awareness is the *tri-kāya* doctrine of the Buddha noted above (the apparitional body, the enjoyment body, and the *dharma* body), a doctrine that was put into its systematic, developed form by Yogācāra thinkers.

The special characteristics of Yogācāra are its emphasis on meditation and a broadly psychological analysis. This contrasts with the other great Mahāyāna system, Mādhyamika, where the emphasis is on logical analysis and dialectic.

This consciousness-oriented school of thought was represented in China primarily by the Fa-hsiang (or Dharmalakṣaṇa; also Wei-shih) school, called Hossō in Japan. The basic teachings of Yogācāra became known in China primarily through the work of Paramārtha, a 6th-century Indian missionary-translator. His translation of the *Mahāyāna-saṃparigraha-śāstra* provided the foundation for the She-lun school, which preceded the Fa-hsiang school as the vehicle of Yogācāra thought in China. Fa-hsiang was founded by Hsüan-tsang, the 7th-century Chinese pilgrim-translator, and his main disciple, K'uei-chi. Hsüan-tsang went to India and studied the doctrines derived from Dharmapāla (d. 507) and taught at the Vijñānavāda centre at Valabhī. When he returned to China he translated Dharmapāla's *Vijñapti-mātratā-siddhi* and many other works. His teachings followed mainly the line of Dharmapāla but also included the ideas of other Indian teachers such as Dignāga and Sthiramati. They were expressed systematically in his works *Fa-yuan-i-lin-chang* and *Wei shih-shu chi,* the basic texts of the Fa-hsiang school.

Characteristics of *dharma*s (*dharmalakṣaṇa*)

Fa-hsiang is the Chinese translation of the Sanskrit *dharmalakṣaṇa* ("characteristic of *dharma*"), referring to the school's basic emphasis on the peculiar characteristics (*dharmalakṣaṇa*) of the *dharma*s that make up the world which appears in human ideation. The connection of this so-called idealist school with the "realist" Abhidharma-Kusha school (see above *Sarvāstivāda* [*P'i-t'an, Chü-she/Kusha*]) is evident, though many new elements are introduced. According to Fa-hsiang teaching, there are five categories of *dharmas*: (1) 8 mental *dharma*s (*cittadharma*), comprising the 5 sense consciousnesses, cognition, the cognitive faculty, and the store consciousness, (2) 51 mental functions or capacities, dispositions, and activities (*caitaśikadharma*), (3) 11 elements concerned with material forms or appearances (*rūpa-dharma*), (4) 24 things, situations, and processes not associated with the mind—*e.g.,* time, becoming (*cittaviprayuktasaṃskāra*), and (5) 6 noncreated or nonconditioned elements (*asaṃskṛtadharma*)—*e.g.,* space or "suchness" (tathata).

Hsüan-tsang's work *Ch'eng wei-shih lun* explained how there can be a common empirical world for different individuals who construct or ideate particular objects and who possess distinct bodies and sensory systems. According to Hsüan-tsang, the universal "seeds" in the store consciousness account for the common appearance of things and particular "seeds" for the differences.

Fa-hsiang was brought to Japan on various occasions, according to the traditional accounts: first by Dōshō, a Japanese priest who visited China, studied under Hsüan-tsang, and established the teaching (now called Hossō) at Gangō-ji monastery, and then by other priests, Japanese and Korean, who studied in China under Hsüan-tsang, K'uei-chi, or their disciples. Thus, the Japanese claim to have received the Hossō teaching in a direct line from its originators, and it continues to have a living and significant role in Japanese Buddhism.

Identity of proximate and ultimate

Avataṃsaka (Hua-yen/Kegon). In contrast with the Fa-hsiang (Hossō) concentration on the specific differentiating characteristics of things and its separation of facts and ultimate principles, the Avataṃsaka school (called Hua-yen in China, Kegon in Japan) stressed the sameness of things, the presence of absolute reality in them, and the identity of facts and ultimate principles. It took its name from the *Mahāvaipulya-Buddhāvataṃsaka-sūtra* ("The Great and Vast Buddha Garland Sutra"), often called simply the *Avataṃsaka-sūtra* ("Wreath Sutra" or "Garland Sutra").

According to legend, the *Avataṃsaka-sūtra* was first preached by the buddha Vairocana expressing the perfect truth revealed in his Enlightenment but then kept secret when it proved incomprehensible to his hearers and replaced with easier doctrines. The sutra tells of the pilgrimage of a young man in a quest to realize *dharma-dhātu* ("totality" or "universal principle"). Extant are three Chinese versions and one Sanskrit original (the *Gaṇḍavyūha*) that contains the last section only. There is no trace of an Indian sectarian development, and the school is known only in its Chinese and Japanese forms.

The school was preceded in China by the Ti-lun school, based on a translation (early 6th century) of Vasubandhu's *Daśabhūmika-sūtra,* concerning the 10 stages of a bodhisattva on the way to buddhahood; since this work was related to the *Avataṃsaka-sūtra,* the Ti-lun adherents readily joined the Hua-yen school established in the late 6th century (?) by Tu-shun (Fa-shun), the first patriarch (d. 640). Fa-tsang (or Hsien-shou), the third patriarch (d. 712), is considered the real founder because he systematized the teachings; hence, it is also called the Hsien-shou school. His student, Ch'eng-kuan (of Ch'ing-liang monastery; d. *c.* 820 or *c.* 838), wrote famous commentaries on the *Avataṃsaka-sūtra.* After the death of the fifth patriarch, Tsung-mi, in 841, Hua-yen declined during the general suppression of Buddhism that ensued in China. But it greatly influenced the development of Neo-Confucianism (a significant movement in Chinese thought beginning in the 11th century) and is regarded by many as the most highly developed form of Chinese Buddhist thought. It was brought into Japan by pupils of Fa-tsang and an Avataṃsaka missionary from central India during the period from about 725 to 740 and began a vital and important development there that has continued down to the present day.

Causation by "totality" (*dharma-dhātu*)

The most significant doctrine associated with this school is the theory of causation by *dharma-dhātu—i.e.,* that all of the elements arise simultaneously, that the whole of things creates itself, that ultimate principles and concrete manifestations are interfused, and that the manifestations are mutually identical. Thus, in Fa-tsang's example of the golden lion in the imperial palace, gold is the essential nature or principle (*li*) and lion is the particular manifestation or form (*shih*); moreover, as gold, each part or particle expresses the whole lion and is identical with every other part or particle. When this model is applied to the universe, it suggests that all phenomena are the expressions of the ultimate suchness or voidness, while at the same time they retain their phenomenal character; each phenomenon is both "all" and "one." All the constituents of the world (the *dharmas*) are interdependent, cannot exist independently, and each of them possesses a sixfold nature: universality, speciality, similarity, diversity, integration, and differentiation.

The ideal is a harmonious totality of things encountered in the perfectly enlightened buddha. The buddha-nature is present potentially in all things. There are an infinite number of buddhas and buddha realms. There are myriads of buddhas in every grain of sand and a buddha realm at the tip of a hair.

The universe is fourfold: a world of factual, practical reality; a world of principle or theory; a world of principle and facts harmonized; and a world of factual realities interwoven and mutually identified. The first three aspects are the particular emphases of other Buddhist schools. The fourth aspect—emphasizing the harmonious whole—is the distinctive doctrine that represents the perfect knowledge that was attained by the buddha Vairocana and is communicated in the *Avataṃsaka-sūtra.*

Figure 3: The Japanese priest Myōe (1173–1232), venerated as the restorer of the Kegon sect, depicted at his daily meditation seated in a tree near his monastery. Hanging scroll, light ink and colour on paper, 13th century. In the Kōzan Temple, Kyōto, Japan.

Kozan-ji, Kyoto, Japan

Saddharmapuṇḍarīka (T'ien-t'ai/Tendai). The T'ien-t'ai/Tendai school is one of the most important developments in Chinese and Japanese Buddhism; it is significant not only for its doctrines, which in many respects are similar to Hua-yen/Kegon, but also for its practical devotional influence. Its doctrines and practices are focused on the Indian or Central Asian *Saddharmapuṇḍarīka-sūtra* ("Lotus of the True Law Sutra"), or *Lotus Sutra.* Also central in the Nichiren school (see below) and recited in Zen temples, this sutra is one of the best known and most popular of Mahāyāna Buddhist texts. (The *Mahāparinirvāṇa* and *Mahāprajñāpāramitā-sūtras* were also important in the development of T'ien-t'ai/Tendai.) The school, which apparently had no separate development as such in India, is sometimes also called Lotus (Fa-hua in Chinese; Hokke in Japanese) but is usually known as T'ien-tai in China and Tendai in Japan, after the mountain in southeastern China where the basic interpretation of the *Lotus Sutra* was first propounded in the 6th century. Prior to this, the original Sanskrit text was studied extensively in China; it was translated into Chinese early in the 5th century by Kumārajīva, and it was taught in North China by the monks and first patriarchs, Hui-wen and Hui-ssū. The latter's student, Chih-i, who settled on Mount T'ien-t'ai and established a famous monastery there, is regarded as the true founder of the school because he propounded the systematic and definitive interpretation of Lotus doctrines. These later became known in Japan, where Saichō (later Dengyō Daishi), a Buddhist priest who studied them first in Japan and then on Mount T'ien-t'ai, founded a new Tendai Lotus Sect (early 9th century) and a monastery on Mount Hiei that became a great centre of Buddhist learning. With Shingon (see below), with which it was closely connected, Tendai became perhaps the most important religious and philosophical influence on the Japanese spirit. Tendai has been markedly syncretistic, striving to include various Buddhist tendencies, from Vinaya to Shingon and Zen, and also Shintō, the indigenous Japanese tradition.

The one way that includes all ways

The *Lotus Sutra*'s main purpose is to establish the one way (or "vehicle" or "career") for attaining salvation (buddhahood). It claims to be the definitive and complete teaching of the Buddha, here presented as a transcendent eternal being, preaching to myriad arhats, gods, bodhisattvas, and other figures using all sorts of sermons, lectures, imaginative parables, and miracles. Religious merit is attributed to preaching, reciting, and hearing the sutra, which is not merely a statement of the nature of things but also a central object of devotion. The three ways of salvation preached by the Buddha are adjusted to the level and situation of the hearers: *śrāvakayāna,* the way of the disciples (*śrāvaka*s), appropriate for becoming an arhat; *pratyeka-buddhayāna,* the way of those who aim at salvation for themselves alone; and *bodhisattvayāna,* the way of those (the bodhisattvas) who, on the point of attaining salvation, give it up to work for the salvation of all other beings. All are forms of the one way, the *buddhayāna.* Being a buddha is the one aim for all.

As systematized in the T'ien-t'ai/Tendai tradition, the Buddha's teachings are divided into five periods. The first is the period immediately following the Buddha's Enlightenment, when, without success, he preached the *Avataṃsaka-sūtra* (or *Hua-yen/Kegon Sutra*). The second is the so-called Deer Park period, when he preached the *Āgama*s (Hīnayāna scriptures) to those with ordinary human capacities. The third is the so-called Fang-teng (broad and equal) period, when he preached the Vaipulya or early Mahāyāna teachings, which were intended for all persons. During the fourth period he preached the Mahāprajñāpāramitā, or Ta-pan-jo-po-lo-mi-to, doctrines concerning absolute voidness and the falsity of all distinctions. Finally, in the fifth and culminating Saddharmapuṇḍarīka and Mahāparinirvāṇa (or Wisdom) period, he taught the identity of contrasts, the unity of the three "vehicles," and the supreme character of the *Lotus Sutra.*

Central to T'ien-t'ai/Tendai doctrine is the threefold truth principle (following Nāgārjuna's [?] commentary on the *Mahāprajñāpāramitā*): (1) that all things are void, without substantial reality, (2) that all things have temporary existence, and (3) that all things are in the mean or middle state, synthesizing voidness and temporary existence, being both at once. The three truths are a harmonious unity, mutually including one another, and the mean or middle truth is equivalent to the absolute suchness. The world of temporary appearances is thus the same as absolute reality.

T'ien-t'ai/Tendai cosmology

T'ien-t'ai/Tendai propounds an elaborate cosmology of 3,000 realms. First are the 10 basic realms, respectively, of buddhas, bodhisattvas, pratyeka buddhas, *śrāvaka*s, heavenly beings, fighting spirits (asuras), human beings, hungry spirits or ghosts (pretas), beasts, and depraved hellish beings. Since each includes the other nine and their characteristics, the 10 are squared to 100 realms. Each of these in turn is characterized by the 10 features of suchness manifested through phenomena: form, nature, substance, power, action, cause, condition, effect, compensation, and ultimacy. Thus the 100 realms in fact constitute 1,000 realms. Furthermore, each of the 1,000 realms comprises three divisions: living beings, space, and the aggregates (skandhas); hence, the whole of things consists of 3,000 realms.

These 3,000 realms interpenetrate one another, are mutually immanent, and are immanent in one moment of thought: "one thought is the three thousand worlds." The universe is not produced by thought or consciousness but is manifest in it, as is the absolute Suchness: hence, the central importance of concentration (*chih*) and insight (*kuan*) that leads to a realization of the unity of things and their manifestation of the ultimate.

Pure Land (Sukhāvatī/Ching-t'u/Jōdo, Shin, and Ji). The main text of the Pure Land schools is the *Sukhāvatīvyūha-sūtra* (*Pure Land Sutra*) written in northern India probably before the beginning of the 2nd century AD. (There are two original versions of the *Sukhāvatīvyūha.* The longer one includes an emphasis on good works; the shorter version emphasizes faith and devotion alone.) This sutra tells of a monk, Dharmākara, who heard the preaching of Lokeśvararāja Buddha many aeons ago and asked to become a buddha. After millions of years of study, Dharmākara promised to fulfill a number of vows if he finally attained buddhahood. He vowed to establish a Pure or Happy Land (Sanskrit: Sukhāvatī; Chinese: Ching-t'u; Japanese: Jōdo), also known as the Western Paradise. In this Pure Land no evil would exist, the people would be long-lived, they would receive whatever they desired, and from there they might attain nirvana. Dharmākara then revealed in a number of vows the means by which this Pure Land can be reached. Several of these vows emphasize meditation and good works on Earth as a prerequisite, but the 18th one (a famous vow in the later development of Pure Land schools) states that, if one merely calls the name of the Buddha at the moment of death, then that person will be reborn in the Pure Land.

Many years after these vows, Dharmākara attained buddahood and now sits in his Pure Land fulfilling his promises of helping humans achieve salvation. Here he is known as the Buddha of Unlimited Light (Sanskrit: Amitābha; Chinese: O-mi-t'o-fo; Japanese: Amida) or the Buddha of Unlimited Lifespan (Amitāyus). He is flanked by Avalokiteśvara (Chinese: Kuan-yin; Japanese: Kannon) on his left and Mahāsthāmaprāpta on his right, who assist Amitābha in bringing the faithful to his Pure Land.

Transmission to Japan

The Amitābhist doctrine spread from India to China by the 3rd century AD, where, through the work of five patriarchs, a sect based on it gradually became the most popular form of Buddhism. The sect was then transferred to Japan by the followers of the Tendai school, who attempted to weld the many sects of Buddhism into one system. By the 13th century AD the Pure Land sect had separated from the Tendai school and spread among the common people of Japan through the work of two outstanding figures, Hōnen and Shinran.

The basic doctrines of the Pure Land sects differ considerably from the doctrines of the early Buddhists. The Pure Land's leaders have generally taught that a person reaches salvation from this Earth not by individual effort or the accumulation of merit but through faith in the grace of the Buddha Amitābha. The main practice of those who follow the Pure Land teachings is not the learning of the texts nor meditation on the Buddha but rather the constant invocation of the name Amitābha. This practice, based on the 18th vow of Dharmākara, the future Amitābha, is called *nien-fo* in Chinese and *nembutsu* in Japanese. Furthermore, in Pure Land Buddhism, the attainment of nirvana is not the primary goal; it is rather to become reborn in the Pure Land of Amitābha.

These doctrines and the practice of invoking the name Amitābha gained great popularity in China and Japan, where it was believed that the world had reached a degenerate period in which the Buddhist doctrines were no longer clear and humans no longer possessed the purity of heart or determination to attain salvation by self-endeavour. Therefore, all people of every section of society could only hope to be saved by the grace of Amitābha. As the Pure Land sect spread from India to China and then to Japan, this doctrine of grace became more and more radical until individual actions were said to play no part in the attainment of salvation.

There is little available data on the practices of the Amitābhist believers in India, but scholars hold that while *nembutsu* was used, the main emphasis was upon meditation and worship of the Buddha. In China this stress on meditation and rites weakened, as indicated in the teachings of three important Pure Land patriarchs, T'an-luan, Tao-ch'o, and Shan-tao, who lived during the 6th–7th century.

T'an-luan was originally a follower of Taoism, who, while searching for the elixir of immortality, was converted to the Pure Land doctrine by an Indian monk. Dedicating his life to the spread of this doctrine, T'an-luan preached the invocation of the name Amitābha and declared that even evil persons were eligible for the Pure Land if they sincerely uttered the *nembutsu.* He warned, however, that the lowest hell awaited those who reviled the Buddhist *dharma.* T'an-luan was followed by Tao-ch'o, who argued that in this degenerate age people must take the "easy path" to salvation of complete trust in Amitābha, for they no longer possessed the capacity to follow the more difficult path of the saints. His disciple, Shan-tao, believed by the Japanese Pure Land sect to be the incarnation of Amida, was primarily responsible for shaping the doctrines of the later forms of Pure Land Buddhism. This evangelist distributed many copies of the *Pure Land Sutra* and wrote a famous commentary in which he taught that rebirth in the Western Paradise is effected primarily by *nembutsu.* This must be supplemented, however, by the chanting of sutras, meditation on the Buddha, worshiping of buddha images, and singing his praises.

Salvation
by Amida's
grace:
Hōnen and
Shinran

The work of Shan-tao inspired Hōnen, the founder of the Pure Land sect (Jōdo-shū) in Japan, to declare that in this evil period people must put complete faith in the saving grace of Amida and constantly invoke his name. Hōnen, who was well versed in Buddhist knowledge, wrote a treatise (*Senchaku hongan nembutsu-shū;* 1198) expressing his beliefs. While this treatise proved popular among the common people, as were his teachings generally, it was burned by the monks of Mount Hiei and his teachings were vigorously opposed by the established Buddhist priesthood.

One of Hōnen's disciples, Shinran, who was exiled at the same time, was the founder of a more radical sect named the True Pure Land sect (Jōdo Shinshū, or Shin). Shinran married, with Hōnen's consent, proving that one need not be a monk to attain the Pure Land; and he popularized his doctrines by preaching in Japanese villages. In his teachings he rejected all sutras except the *Pure Land Sutra* and rejected the vows of Dharmākara in the *Pure Land Sutra* that stress individual merit. Basing his doctrines on the 18th vow, Shinran discouraged any attempt to accumulate merit, for he felt that this stood in the way of absolute faith and dependence on Amida. Furthermore, he rejected Hōnen's practice of continual invocation of Amida, believing that the faithful need only say the *nembutsu* once in order to attain salvation. Any repetition after this *nembutsu* must be seen as praise of Amida and not as bringing merit or affecting one's salvation. Thus, with Shinran, the doctrine of grace gained total ascendancy. A third Pure Land sect grew up around the itinerant teacher Ippen. He traveled throughout Japan, advocating the chanting of Amida's name at set intervals throughout the day; hence, his school was called the Ji ("Times") sect, or Jishū.

Music, dance, and drama have been important forms of expression of Shin. Since the late 19th century it has engaged in extensive educational and social welfare programs and has played a significant role in Japanese life. It is the largest single Buddhist sect in Japan.

Nichiren. The indigenous Japanese Nichiren school is related both to the *Lotus Sutra* and Pure Land schools, for it, too, is centred on the "Lotus of the True Law" and also emphasizes fervent faith and the repetition of a key phrase. Hence it has been aptly called "Lotus-pietism" by a famous scholar in Japanese Buddhism. Its distinctiveness is rooted in the extraordinary personality and character of its founder; significantly, it is named after a man, a historical person, not after a book or a doctrine. Nichiren (1222–82), the son of a poor fisherman, became a monk at an early age and studied at Mount Hiei, the centre of the Tendai school. Nichiren was frustrated, however, by the many paths of Buddhism promising salvation and left Mount Hiei for 10 years to search for the true path. When he emerged from his independent studies he taught that the *Lotus Sutra* (*Saddharmapuṇḍarīka-sūtra*) contains the final and supreme teaching of the Buddha Śākyamuni and offers the only true way to salvation. According to Nichiren's interpretation of this sutra, the three forms of the Buddha—the universal or law body (*dharma-kāya*), the enjoyment body (*sambhoga-kāya*), and the phenomenal body (*nirmāṇa-kāya*)—should be granted equal respect, as they are important aspects of the Buddha Śākyamuni. Following the teachings of Chih'i, the Chinese founder of T'ien-t'ai/Tendai, that the *Lotus Sutra* is the essence of Buddhism, Nichiren held that this same buddha-nature was possessed by all people and could be realized only by proper worship of the *Lotus Sutra.* Furthermore, Nichiren felt that his time, which was marked by political upheaval and unrest, was the period of degeneration known in the *Lotus Sutra* as the time of the latter-day *dharma* (*Mappō*). During this time the purity of the Buddhist doctrines could be kept only by the bodhisattvas. Nichiren identified himself as an incarnation of several of these bodhisattvas, especially the bodhisattva of supreme conduct (Viśiṣṭacāritra; Japanese: Jōgyōbosatsu), and believed that his mission was to propagate the true teachings of the *Lotus Sutra* in Japan, which he felt would become the new repository of the Buddhist *dharma.*

Importance
of the
*Lotus
Sutra*

In attempting to guide Japan to the Buddhist *dharma* as he interpreted it, Nichiren drew great criticism for his strong-willed and uncompromising attitude. In one treatise Nichiren wrote that the unrest in Japan was caused by the chaotic state of religious beliefs, a condition that could be corrected only by adopting the teachings of the *Lotus Sutra.* He taught that if people turned to this sutra, they would realize their true buddha-nature, perceive that suffering is illusion, and see that this world is a paradise. If human beings—*i.e.,* the Japanese—did not follow the teachings of the *Lotus Sutra,* however, natural disasters and invasions would result. With firm confidence in the righteousness of his cause, Nichiren attacked the Shingon and Amida sects of Buddhism for neglecting Śākyamuni, the true Buddha of the *Lotus Sutra;* and he attacked Zen for placing stress only upon Śākyamuni's historical form. These sharp criticisms led Nichiren to be exiled twice and almost brought his execution, from which he was—according to his account and the belief of his adherents—miraculously saved.

Nichiren advocated two main religious practices based on his interpretation of the *Lotus Sutra.* The first is the worship of the *honzon* (or *gohonzon*), a mandala (symbolic diagram) designed by Nichiren, representing the buddha-nature that is in all humans, as well as the three forms of the Buddha Śākyamuni. The second is the repetition of the phrase *namu Myōhōrenge-kyō* (salutation to the *Lotus Sutra*), a practice called *daimoku* ("sacred title"), as the affirmation of the devotee's belief in the teaching and efficacy of the *Lotus Sutra.* This repetition was not only to be done orally but in every action of the true believer. Nichiren also taught that there should be a sacred place of ordination (*kaidan*) where the believer could receive training in the doctrines of the *Lotus Sutra* in order that he might keep the true spirit of this document. This sacred place might be seen as wherever the believer in the *Lotus Sutra* lives, for there is the Buddhist truth. The *honzon, daimoku,* and *kaidan* constitute "the three great secret laws" (or "mysteries") that are regarded as the essential teaching of Nichiren.

The great
secret laws

Nichiren's fervent faith brought him wide fame and many devotees, and at his death he chose six disciples to continue his work. This sect was known simply as Nichiren-shū (Sect of Nichiren). It still controls the main temple founded by Nichiren at Mount Minobu. One of his disciples, Nikkō, however, soon began another sect known as the Nichiren-shō-shū (True Sect of Nichiren), which taught that Nichiren, not Śākyamuni, was the saviour and that the mandala painted by Nichiren was alone efficacious in saving mankind. In the 20th century these sects have gained many devotees.

Within the Nichiren-shū the Reiyū-kai (Association of the Friend of the Spirit) arose in 1925. This sect, which preaches a combination of ancestor cult and the doctrines of Nichiren, places faith not in the Buddha or in bodhisattvas but in the mandala, in which all saving power is concentrated. The Risshō-Kosei-Kai (Society for Establishing Righteousness and Friendly Relations) split from Reiyū-kai in 1938. This sect teaches the recitation of the *daimoku* as an affirmation of faith in the teaching of the *Lotus Sutra,* and the worship of the Buddha Śākyamuni. Like Reiyū-kai, it also allows the veneration of ancestral spirits.

Risshō-Kosei-Kai gained many converts after World War II, but its success was soon eclipsed by Sōka-gakkai, the lay movement of Nichiren-shō-shū. Sōka-gakkai ("Value Creation Society") was founded by Makiguchi Tsunesaburo in 1930, dedicated to educational research and the extension of Nichiren-shō-shū. Its founder insisted on the practical values of worldly gain and satisfaction as well as the attainment of goodness and beauty; he taught that Nichiren was to be worshiped as the True Buddha predicted in the *Lotus Sutra.* The members also fervently practice *daimoku* and worship the *honzon* as the repository of magical power. After World War II, Sōka-gakkai, under the leadership of Toda Jōsei, grew rapidly through a technique of evangelism called *shakubuku* ("break and subdue"), in which the resistance of the other person is destroyed by forceful argument. A zealous missionary movement, it has spread to many countries, including the United States. Thus, Nichiren's teaching and personality are still strong influences today.

Dhyāna (Ch'an/Zen). The Dhyāna (Chinese: Ch'an; Japanese: Zen) school of Buddhism emphasizes meditation as the way to immediate awareness of ultimate reality, an important practice of Buddhism from its origin in India, and derives its name from the Sanskrit term for meditation, *dhyāna.* The meditative emphasis in other schools of Indian origin, such as Yogācāra, has been noted above. Ch'an, with its special training techniques and doctrines and its Taoist influence, however, is generally considered a specifically Chinese product.

The enlightenment experience: Satori

Scholars point out that 4th–5th-century Chinese Buddhist monks, such as Hui-yüan and Seng-chao, were teaching doctrines and practices similar to those of the Ch'an school before the traditional date of its arrival in China, but standard Chinese texts name a South Indian monk, Bodhidharma, who came to China about AD 520, as its founder. Bodhidharma is held by Ch'an devotees to be the 28th patriarch of the Indian meditational school, which began with the monk Kāśyapa, to whom the Buddha Śākyamuni revealed his supreme teaching. This teaching is found in the *Laṅkāvatāra-sūtra,* which relates that all beings possess a buddha-nature, often equated with *śūnya* (the void) in Ch'an, and that realization of this fact is enlightenment (Chinese: Wu; Japanese: Satori). The truly enlightened one cannot explain this ultimate truth or reality, for it is beyond the ordinary duality of subject and object. Books, words, concepts, and teachers cannot convey it directly. It must be realized in immediate personal experience.

Bodhidharma, who came to be known as the first patriarch of Ch'an in China, passed his mantle to Hui-k'o; and this line of transmission continued to the fifth patriarch, Heng-jen. After his death a schism occurred between the adherents of the Northern school founded by Shen-hsui, which held that enlightenment must be attained gradually, and the Southern school of Hui-neng, which taught that true wisdom, as undifferentiated, must be attained suddenly and spontaneously. Furthermore, Hui-neng's Southern school tended to neglect rituals and literature and to rely on teaching passed from master to pupil, adopting as well an iconoclastic attitude toward the Buddha. It was reasoned that, if all things contain the buddha-nature, then the Buddha could rightfully be equated with a dung heap. Eventually the Southern school won out, and its victory is attested in the standard Chinese Ch'an texts, which name Hui-neng as the true and only sixth patriarch, as opposed to the counterclaim of Shen-hsui and the Northern school. Hui-neng's *Platform Scripture* (Chinese: *T'an Ching*) became a key text of the Ch'an school.

Two branches of Ch'an Buddhism developed from the Southern school in the 9th century: Lin-chi (Japanese: Rinzai) and Ts'ao-tung (Japanese: Sōtō). The former relied heavily on the *kung-an* (Japanese: *kōan*), a paradoxical question or aphorism that aimed at inducing the student to realize that all conceptualization is wrong, thus leading him to enlightenment. The *kung-an* were often accompanied by shouts and slaps from the master to provoke anxiety and, from this, an instant realization of the truth. The Ts'ao-tung/Sōtō school emphasized the practice of "silent illumination" or "just sitting" (Chinese: *tso-ch'an;* Japanese: *zazen*). This consisted of sitting in silent meditation under the direction of a master and—in that context—of purging the mind of all notions and concepts.

Both schools followed the doctrine of Huai-hai, who taught that a monk who would not work should not eat and that the work (as well as everything else) should be done with spontaneity and naturalness. The emphasis on work made the Ch'an schools self-sufficient and helped to save them from the worst effects of the purge of supposedly parasitic Buddhist monks by the government in 845. The emphasis on spontaneity and naturalness stimulated the development of a Ch'an aesthetic that had a profound influence on later Chinese painting and writing. The relative success the Ch'an tradition was able to achieve in the subsequent religious history of China is demonstrated by the fact that virtually all Chinese monks eventually came to belong to one of the two Ch'an lineages.

Influence of Zen

Ch'an (Zen) Buddhism was introduced into Japan as early as the 7th century, but it did not flower until the 12th century, in the work of two monks, Eisai and Dōgen. Eisai, founder of the Rinzai school in the 12th century, was a Tendai monk who wished to restore pure Buddhism to Japan and with that aim visited China. When he returned, he taught a strict meditational system based on the use of the *kōan* phrases; however, this was taught as only one element in the Tendai system. Unlike the Ch'an schools, Eisai also taught that Zen should defend the state and could observe ceremonial rules and offer prayers and incantations. These teachings influenced the warrior class and led to a Zen influence over the martial arts of archery and swordsmanship. Zen influence can also be seen in the nō play, poetry, flower arrangement, and the tea ceremony, all of which stress grace and spontaneity.

Dōgen, who established the Sōtō school in Japan in the 13th century, joined the Tendai monastery of Mount Hiei at an early age, after the death of his mother and father had dramatically taught him the transitoriness of life. Searching for the true path of Buddhism, he, like Eisai, journeyed to China, where he fell under the influence of a Chinese Ch'an master. Upon his return to Japan, he taught a discipline of "sitting straight" (*zazen*) without any effort being directed toward achieving enlightenment. Unlike many of his Chinese counterparts, however, Dōgen studied scriptures and criticized those who did not.

The two Zen sects founded by Eisai and Dōgen have deeply influenced Japanese culture, and they continue to play a very significant role in contemporary Japan. By the mid-20th century Zen had become perhaps the best known of the Buddhist schools in the Western world.

(Gi.T./F.E.R.)

ESOTERIC BUDDHISM

Mystical practices and esoteric sects are to be found in all forms of Buddhism. In the course of history the mystical tendency that had suffused Buddhism as an aspect or element of Indian religion as a whole became increasingly pronounced. Following the codification of early Buddhism in the Theravāda canon and the subsequent emergence of Mahāyāna about the 1st century AD, this element began to precipitate into discrete schools of thought.

Buddhist mysticism (including the philosophical school of Ch'an, above), like other forms of mysticism, insists on the ineffability of the mystical experience, because it defies expression in terms that are intelligible to anyone who has not had analogous experience. The knowledge involved is never merely intellectual but is a kind of felt knowledge in which things are seen in a different perspective and take on a new significance. Related to its ineffability is the timeless quality of this experience, which means that the mystic seems to be outside time and space, oblivious to his surroundings and the passage of time.

The earliest Buddhist mysticism was concerned with the emptying of subjective being, considered to be the greatest obstacle to the individual's spiritual growth. This passing into a new dimension of reality was described in terms of a flame going out: it was merely extinguished; it could not be said that it had gone somewhere. In this emptying process the limits that constitute the individual's being, as defined by an analysis of both physical and mental components, were transcended, although all these components were said to be retained. The experience of this new dimension of reality was a vision in marked contrast to normal perception—a vision that went far beyond the reach of "mere logic."

While early Buddhism (as preserved in Theravāda) was analytical in its attempt to free reality from the imposition of subjectivity, Mahāyāna continued the analytical process by extending it to objective reality. In its rejection of subjectivism and objectivism, it emphasized the nature of reality-as-such, which was experienced in enlightenment (bodhi). While the various philosophical trends associated with the emergence and development of Mahāyāna dealt with the intellectual problem of reality, shorn of all its positive and negative qualifications, the tantras ("treatises"), which form the distinctive literature of Esoteric Buddhism, dealt with the existential problem of how it is or feels to attain the highest goal.

Vajrayāna Buddhism in India. *Origins.* Vajrayāna ("Diamond Vehicle") or Mantrayāna ("Path of the Sacred Formulas"), also known as Tantric Buddhism, first gained prominence in various parts of India and Sri Lanka. Scholars infer that, because of the esoteric nature of Tantric practice and doctrine, this school might have been developing quietly from the 2nd or 4th century AD, when Buddhist tradition associates Nāgārjuna or Asaṅga with its origins. Although a modified version of Vajrayāna Buddhism, apparently without sexoyogic practices, spread to China and then to Japan, where it became known as Shingon, most scholars associate the Vajrayāna tradition primarily with India and Tibet.

Identity and unity of opposites

Although Vajrayāna texts describe numerous yogic or contemplative stages that an aspirant must experience before achieving enlightenment, rather than elaborating doctrines, they hold the Mahāyāna identification of nirvana and samsara as a basic truth. Moreover, Vajrayāna maintains that nirvana as *śūnyatā* (voidness) is one side of a polarity that must be complemented by karuna (compassion of the bodhisattva). *Śūnyatā* is seen as passive wisdom (prajna) that possesses an absolutely indestructible or diamondlike (*vajra*) nature beyond all duality, whereas karuna is the means (*upāya*) or dynamic aspect of the world. Enlightenment arises when these seeming opposites are realized to be in truth one. This realization, which is known experientially and not through a purely cognitive process, is portrayed in some types of Vajrayāna imagery and practice as the union of the passive female deity, which signifies wisdom or voidness, with the dynamic male, signifying compassion without attachment. Such a union, called *yab-yum* ("father-mother") in Tibetan, is not a satisfaction of physical impulses but a symbol of the unity of opposites that brings the "great bliss," or enlightenment.

Adherents of the Vajrayāna tradition believe that as all things are in truth of one nature—the void—the physical-mental processes can be used as a vehicle for enlightenment. In the *Kālacakra Tantra* it is written that the Buddha taught that in this age of degeneration enlightenment must be achieved through one's own body, which contains the whole cosmos. This doctrine is taught in all the tantras. Vajrayāna specialists warn, however, that in order to use correctly the body's processes to achieve an identification of the void with compassion, the aspirant must follow absolutely the instructions of a master or teacher who has been initiated into the mysteries. Such a master alone can direct every step so that the pupil learns to control mental and physical processes instead of being dominated by them. Therefore, the first step toward enlightenment in Vajrayāna practice is the undergoing of initiation by a master.

Figure 4: Tibetan Buddhist monk reading with handbell in Lamayuru Monastery, Ladakh, India.

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The master first endeavours to direct the student to compassion through meditation on the transitoriness of life, the relation of cause and effect of one's actions, and the general suffering of humanity. After this sympathy for the suffering of humanity is aroused, the master guides his pupil in yogic, or contemplative, exercises that help to produce inner experiences corresponding to the various stages of spiritual growth. This process of advancement toward enlightenment involves the identification of the initiate with gods or goddesses that represent various cosmic forces. These gods are first visualized with the help of mudras (meditative gestures and postures), mantras (sacred syllables and phrases), and icons, all of which are believed to possess the essence of the divinities to be invoked. The icons are portrayed in a mandala, a sacred design that represents the universe as an aid to meditation. After this visualization, the initiate identifies with the divinities and finds that each in turn is *śūnyatā,* or voidness.

The four stages

The culmination of this process, called *vajrasattva yoga,* gives the initiate a diamondlike body beyond all duality. Four stages in the process are described in four different groups of tantras, the *Kriyā-tantra, Caryā-tantra, Yoga-tantra,* and *Anuttarayoga-tantra.* These four stages are likened to the fourfold phases of courtship: the exchange of glances, a pleasing or encouraging smile, the holding of hands, and consummation in the sexual act. The first stage involves external ritual acts, whereas the second combines these outward acts with contemplation. The third stage involves only contemplation, and the fourth is the unification of all dualities in the sexual act, symbolically or effectively. This last stage, however, is divided into two phases. The first involves the use by the initiate of controlled imagination, which allows him to experience the union on an ideational level. The second phase is the *maithuna,* or sexual coupling. This act, however, cannot be construed as an ordinary physical mating, because the initiate has already realized the voidness of all things, allowing him to act with perfect control over his emotions and without attachment. Whereas the ordinary sexual act gives rise to only momentary pleasure, this *maithuna* is considered to be an appropriate technique for attaining enlightenment and eternal bliss.

These Vajrayāna practices have been unjustly condemned as a degeneration of Buddhism by those who do not look beyond the surface. It is quite easy to misinterpret the *Guhyasamāja-tantra* when it states that adultery and eating of human flesh are actions of the bodhisattva if one does not realize that this imagery points to the belief that voidness alone exists, beyond good or evil, or that the initiate must act only with compassion for the benefit of the salvation of the world. Once the true depths of Vajrayāna doctrines and practices are perceived, this school can be designated as a development of Buddhist thought that emphasizes the attainment of enlightenment through a graduated process of meditation under the direction of an initiated teacher.

Vajrayāna literature. While the sutras can be said to represent the theoretical and speculative aspect of Buddhism, the tantras, written in a highly figurative language, express Buddhism as individually lived. The tantras are essentially individually oriented works that apply to larger groups of persons because of the similarity of the experiences described in them. The individual spiritual development occurs through symbols that must not be reduced to signs; a symbol always points beyond itself. Because of this symbolic character, the tantras have usually been kept secret, and a literalist interpretation of such texts has usually failed to make any sense out of them.

The *Guhyasamāja-tantra* ("Treatise on the Sum Total of Mysteries"), also known as the *Tathāgataguhyaka* ("The Mystery of Tathāgatahood [Buddhahood]"), is the earliest known written tantra. It is by tradition ascribed to the renowned Indian scholar Asaṅga (*c.* 4th century AD), the propounder of the Yogācāra philosophy. Usually the tantras do not give an explanation of the technical or symbolic terms, as this explanation is left to the teacher, but the *Guhyasamāja-tantra* devotes a very long chapter to the elucidation of these terms.

Polarity symbolism

An important feature of all tantras is a polarity symbolism, which on the physical level appears as the union of male and female; on the ethical level it appears as the union of beneficial activity and an appreciation of what there is as it is; and on the philosophical level it appears as the synthesis of absolute reality and absolute compassion. The richness of this symbolism is already indicated in the opening of the *Guhyasamāja,* where the absolute, which is depicted as a polarity, manifests itself in various mandalas (circular diagrams that have both a psychological and cosmic reference), each related to one of the celestial, meditational buddhas—Akṣobhya, Vairocana, Ratnasambhava and Amitābha, and Amoghasiddhi—each of whom again represents a polarity, portrayed in iconographic works through their union with their female consorts.

The ideas and symbols presented in the *Guhyasamāja-tantra* became in the course of time more and more clearly elaborated. Because the tantras reflect an individual process of growth, the centre toward which this process gravitates, and from which it is also fed, appears in various symbols given various designations. Thus, there is the *Hevajra-tantra,* in which the sustaining life force is called *Hevajra,* and the *Mahāvairocana-tantra,* in which it is called Mahāvairocana ("The Great Resplendent One").

In view of the fact that the tantras may emphasize either "beneficial activity" or "appreciative awareness" or their "unity," the Tantric literature has been divided into the so-called *Father Tantra* (emphasizing activity), the *Mother Tantra* (emphasizing appreciation), and the *Nondual Tantra* (dealing with both aspects unitively). Almost all these works have been lost in their original Sanskrit versions, but their influence is noticeable in such works as *Jñānasiddhi* ("Attainment of Knowledge") by the great Vajrayāna teacher Indrabhūti (*c.* 687–717), *Prajñopāyaviniścayasiddhi* ("The Realization of the Certitude of Appreciative Awareness and Ethical Action") by the 8th-century writer Anaṅgavajra, and the songs of the 84 *mahāsiddhas* ("masters of miraculous powers," who were considered to have attained the Vajrāyana goal). One of the last Sanskrit works to have been written in Central Asia was the *Kālacakra-tantra.* Its penetration into India may be dated AD 966. The central theme is the Ādi-Buddha—primeval buddhahood—manifesting itself as a continuum of time (*kāla*) and space (*cakra*).

Vajrayāna schools in Tibet. During the period when Tibet was being converted to Buddhism (7th to 11th centuries), the most dynamic form of Buddhism in India was Vajrayāna. Thus, it was this Buddhist tradition that became established in Tibet. Though it is probable that some form of Vajrayāna played an important role in the original conversion of Tibet (7th to 9th centuries), little is known about the details. Beginning in the 11th century, however, it is possible to identify several different schools.

Like most Buddhist schools, those that developed within the Tibetan tradition were highly permeable associations that encompassed a number of different and often competing lineages. This having been said, it is still possible to single out several groupings.

Rnying-ma-pa. Among the various Vajrayāna schools of Tibet and neighbouring regions, the Rnying-ma-pa claims to preserve most purely the spirit of the teachings of the 8th-century Indian miracle worker Padmasambhava. It makes fuller use than any other school of a group of "discovered" texts, said to have been hidden since the period of persecution that began in Tibet in the first half of the 9th century. (In the Rnying-ma-pa tradition the notion of "hidden treasure" has strong spiritual as well as historical overtones.) The discoveries of hidden texts associated with Padmasambhava began to occur in the 11th century and have continued until very recently.

Discovered texts

The Rnying-ma-pa order divides Buddhist revelation into nine progressively superior groups; it also subdivides the tantras in a manner somewhat different from the way they are divided in other Vajrayāna schools. Six groups of tantras are enumerated: (1) *Kriyā,* or ritual, (2) *Upayoga,* which involves the convergence of the two truths and meditation on the pentad of buddhas, (3) *Yoga,* which involves the evocation of the god, the identification of the self with the god, and meditation on the mandala (ritual drawing), (4) *Mahāyoga,* which involves meditation on the factors of human consciousness (skandhas) as divine forms, (5) *Anuyoga,* which involves secret initiation into the presence of the god and his consort and meditation on "voidness" in order to destroy the illusory nature of things, and (6) *Atiyoga,* which involves meditation on the union of the god and his consort, leading to the experience of bliss. Those initiated into the *Kriyātantra* can attain buddhahood after seven lives, the *Upayoga* after five lives, the *Yoga* after three lives, the *Mahāyoga* in the next existence, the *Anuyoga* at death, and the *Atiyoga* in the present existence.

Among the exponents of the Rnying-ma-pa tradition, Klong-chen rab-'byams-pa (1308–63), who wrote the *Klong-chen-mdzod-bdun* ("Seven Treasures of Klong-chen"), is one of the most profound Vajrayāna thinkers. More recently, Mi-'pham of Khams (1846–1914) wrote important Vajrayāna commentaries on the canonical texts.

Sa-skya-pa, Bka'-brgyud-pa, and related schools. Several Tibetan schools that came into being during the 11th and 12th centuries traced their lineage back to particular Vajrayāna saints who had lived in India some centuries earlier. Among these the Sa-skya-pa and the Bka'-brgyud-pa orders were especially prominent, and during the course of Tibetan history they gave rise to many other orders.

The Sa-skya-pa order traces its lineage back to an Indian *mahāsiddha* named Virupa. The order's founder was a Tibetan named 'Brog-mi (992–1072) who went to India, received training in the Vajrayāna, and translated the *Hevajra-tantra* into Tibetan; the order places a strong emphasis on this tantra. The Sa-skya-pa system is also called *lam-'bras* ("the path and its fruit").

The Sa-skya-pa order produced many great translators, and its scholars also contributed original works on Vajrayāna philosophy and linguistics. On the ecclesiastical and political level the order sometimes exerted considerable power. During the 13th century, for example, the Sa-

Initiation of Kublai Khan

skya-pa abbot ’Phags-pa (1235–80?) conferred initiation according to the *Hevajra-tantra* on Kublai Khan (founder of the Yüan, or Mongol, dynasty in China) and in turn was appointed *ti-shih* (Chinese: “Imperial Preceptor”) and invested with the authority to govern Tibet, though under the control of the Mongol court.

The Bka’-brgyud-pa school traces its spiritual lineage from the Indian master Tilopa, who transmitted the teachings to the Indian yogi Naropa, the master of Mar-pa, the 11th-century householder-teacher, who was in turn the master of Mi-la-ras-pa (1040–1123). The school has preserved an important collection of songs attributed to Mi-la-ras-pa and a fascinating hagiographic account of his life. Sgam-po-pa (1079–1153), who was his greatest disciple, did much to systematize the teaching and to establish the basis for further development. His most famous work, *Thar-rgyan* (“The Jewel Ornament of Liberation”), is one of the earliest examples of a genre that became extremely important in the later development of the Vajrayāna tradition in Tibet and Mongolia. Known as *Lam Rim* (“Stages on the Path”), this genre presents the whole of Buddhist teachings in terms of gradations in a soteriological process leading to the attainment of buddhahood.

The Bka’-brgyud-pa teachers stressed, among other techniques, the exercises of Haṭha Yoga (a yoga emphasizing breathing and special postures) and posited as the supreme goal the *mahāmudrā* (“the Great Seal”), or the overcoming of dichotomous thought in the emptiness of buddhahood. Frequent reference is made by the Bka’-brgyud-pa to the “Six Teachings of Naropa.” These teachings set forth techniques for attaining enlightenment, either in this life or at the moment of death, that are associated with: (1) self-produced heat (the voluntary raising of the body temperature), (2) the illusory body, (3) dreams, (4) the experience of light, (5) the state of existence intermediate between death and rebirth (Bardo), and (6) the passing over from one existence to another.

Among the many lineages that have developed within the Bka’-brgyud-pa order, the one that is best known today is the Karma-pa (Black Hat) lineage, which has its major centre at the monastery of Mtshur-phu.

The Bka’-gdams-pa and Dge-lugs-pa. The Bka’-gdams-pa school was based on the teachings of Atīśa (an Indian monk who came to Tibet in the 11th century). It was founded by his chief disciple, ’Brom-ston (*c.* 1008–*c.* 1064), who emphasized austere discipline. The school produced the *Bka’-gdams gces-bsdus* (“Collection of the Sayings of the Bka’-gdams-pa Saints”), which preserves the often highly poetic utterances of those close to the founder. The central practice of the school was the purification of the mind, which required the elimination of intellectual and moral blemishes in order to obtain a clear vision of emptiness (*śūnyatā*). The school relied on the *Prajñāpāramitā* and related texts and also made use of mantras. The Bka’-gdams-pa order was absorbed in the 15th century by the Dge-lugs-pa school.

Reforms of Tsong-kha-pa

The Dge-lugs-pa (Gelugpa; the “Virtuous”) represents the reformed sect in Tibet; its members are commonly known as Yellow Hats from the colour of their head cover. Their founder, Tsong-kha-pa, attended the most important schools in 14th-century Tibet, the Sa-skya-pa, Bka’-brgyud-pa, and Bka’-gdams-pa. His own school is considered the continuation of the Bka’-gdams-pa. Tsong-kha-pa was prompted to initiate a reform of monastic discipline by what he considered to be a general laxity of morals, an increasingly less rigorous observance of monastic rules, and the prevalence of deviations in the interpretation of the tantras. He imposed respect for the traditional rules of the *Vinaya* and placed renewed emphasis on dogmatics and on logic as aids to salvation. His treatise, the *Lam-rim chen-mo* (“The Great Gradual Path”), is based on the *Bodhipathapradīpa* by Atīśa. In it Tsong-kha-pa presents a process of mental purification ascending through 10 spiritual levels (*bhūmi*) that lead to buddhahood. The essential points of such a process are the state of quiescence and the state of enhanced vision.

Tsong-kha-pa attributed great importance to the study of logic and instituted regular debates at monasteries. Competing monks sought to reach, by means of formal logic and in the presence of an abbot of great learning, an unassailable conclusion on a chosen topic. Various ranks of monks were established on the basis of examinations, the highest being that of *dge-bshes* (the philosophers).

This insistence on the importance of doctrinal and logical problems did not exclude interest in the tantras, and Tsong-kha-pa’s *Sngags-rim chen-mo* (“The Great Gradual Tantric Path”) deals with Tantric ritual. Tantric initiation, however, was open only to those students who had previously mastered theoretical learning. The literature of the Dge-lugs-pa is enormous, including also the gigantic collections of the Dalai and Panchen lamas, both of whom are members of this school.

The Dge-lugs-pa assert that the nature of the mind element (*sems*) is light, which constitutes the cognitive capacity. The continuum of each person, therefore, is a thinking and luminous energy, which is in either a coarse or subtle state, the latter state being achieved only after purification through meditation and contemplation.

(Gi.T./H.G./F.E.R.)

Esoteric traditions in China and Japan. During the 7th, 8th, and 9th centuries the Vajrayāna forms of Esoteric Buddhism that were developing in India spread to Southeast Asia (for example, to Indonesia, where Esoteric teachings were an important component in the symbolism of the great Buddhist stupa at Borobuḍur) and to East Asia. In East Asia Esoteric Buddhism became established in the Chen-yen (“True Word”) school in China and in the Tendai and Shingon schools in Japan.

The basic Chen-yen texts

Chen-yen. According to the Chen-yen tradition, developed and systematized forms of the Esoteric tradition were first brought from India to China by three missionary monks: Śubhākarasiṃha, Vajrabodhi, and Amoghavajra. Śubhākarasiṃha arrived in China from the famous Indian centre of learning at Nālandā in 716, and he translated into Chinese the *Mahāvairocana-sūtra* and a closely related ritual compendium known as the *Susiddhikāra*. Vajrabodhi and his disciple Amoghavajra arrived in 720 and produced two abridged translations of the *Sarvātathāgatatattvasaṃgraha* (“Symposium of Truth of All the Buddhas”), also known as the *Tattvasaṃgraha*. The *Tattvasaṃgraha* and the *Mahāvairocana-sūtra* became the two basic Chen-yen texts.

In the period of approximately 130 years between the arrival of Śubhākarasiṃha and the great persecution of 845, the Chen-yen school enjoyed amazing success. The tradition represented by Śubhākarasiṃha and the *Mahāvairocana-sūtra* and the tradition represented by Vajrabodhi and the *Tattvasaṃgraha* were melded. Chinese disciples such as Hui-kuo were initiated into a common lineage and contributed to an emerging Chen-yen synthesis. Through a combination of sophisticated doctrinal instruction and the exercise of the miracle-working powers supposedly conferred by the Esoteric rituals, the Chen-yen leaders won the confidence of the court, especially of Emperor Tai-tsung (762–779/780), who rejected Taoism in favour of Buddhism in its Chen-yen form. Chen-yen became a dominant force among the Chinese elite.

After the great persecution of 845, Chen-yen lost its position of prominence in China, but it maintained a certain degree of spiritual vitality and communal visibility on through the Sung dynasty (960–1279). Moreover, the Chen-yen school contributed a variety of elements—particularly ritual elements—that have become persisting threads within the larger fabric of Chinese religion.

Shingon. Though Esoteric Buddhism played a much greater role in China than is usually recognized, it was in Japan that it became most firmly established and exerted its most extensive influence. Esoteric elements, called Taimitsu in Japanese, became and have remained an important element in the Japanese Tendai school (see above). The Tendai school was founded by the monk Saichō (764–822), who studied with Chen-yen and T’ien-tai masters in China. The most systematized and elaborated expressions of the Esoteric tradition, however, were developed in the Shingon school, the Japanese version of Chen-yen.

The founder of the Shingon school in Japan was Kūkai, better known by his posthumous name, Kōbō Daishi (“Great Master Who Understood the *Dharma*”). He was an exceptional scholar, poet, painter, and calligrapher who early in life wrote a treatise comparing Confucian, Taoist, and Buddhist thought, naming the latter as superior. Although he had trained for government service, he experienced a change of heart and became a Buddhist monk. In pursuit of the pure Buddhist doctrine, he, like many great monks of his time, journeyed to China, where he met the Chen-yen master Hui-kuo, who recognized Kūkai's potential and bestowed upon him the teachings of his school. After the death of Hui-kuo, Kūkai returned to Japan, where he received many governmental honours and established a monastery on Mount Kōya as the centre of Shingon Buddhism.

In propagating the teachings of his sect, Kūkai wrote many important texts, including the *Jūjū shinron* ("The Ten Stages of Consciousness") and the *Sokushin-jōbutsugi* ("The Doctrine of Becoming a Buddha with One's Body During One's Earthly Existence"). In the first of these treatises, Kūkai presented a theory of the development of the spiritual life of human beings by placing the teachings of Buddhist schools and several other religions into a hierarchical system. He taught that the first stage of human spiritual development was one in which human beings are completely controlled by their instincts. In the second stage, which Kūkai identifies with Confucian teachings, human beings attempt to live a proper moral existence. The third stage is that of Brahmanism and Taoism, where the individual strives for supernatural powers and heavenly rewards. The fourth and fifth stages of spiritual development are taught by the Hīnayāna schools and are characterized by the striving for self-enlightenment. The next stages, from six to nine, are Mahāyānist paths identified with the teachings of Hossō, Sanron, Tendai, and Kegon, which lead the individual to compassion for others. The zenith of spiritual development is identified by Kūkai with the esoteric teachings of Shingon.

The stages of spiritual development

The Shingon school preached that it possessed the highest and purest doctrine, for its beliefs were not based on the teachings of the historical Buddha, Śākyamuni, who expounded the doctrine with limitations of his audience in mind, but on the timeless and immutable teachings of the Buddha in his *dharma-kāya,* or cosmic body. This Buddha, named Mahāvairocana, was felt to be beyond all earthly dualism and impurity but at the same instant to be within all things as their buddha-nature.

In Shingon the realization that one's own buddha-nature is identical with Mahāvairocana is enlightenment. This enlightenment, as depicted in the title of Kūkai's treatise mentioned above, can be achieved in this world while possessing a human body. In order to achieve this enlightened state, however, one must be given the secret doctrine of Shingon. The gift of this doctrine is handed down to the aspirant only orally by the Shingon master. The truth that the master reveals is founded on the three ritual mysteries of the body, speech, and mind. These mysteries invoke the cosmic forces embodied in the form of buddhas and bodhisattvas with which the aspirant identifies until he can become one with Mahāvairocana. The experience of the mystery of the body involves the use of mudras: various devotional gestures of the hands and fingers in accordance with the buddha to be invoked, postures of meditation, and the handling of such sacred instruments as the *vajra* (diamond) and the lotus. Experiencing the mystery of speech involves the recitation of dharanis or mantras, which are mystical verses and sounds believed to be the essence of the cosmic forces with which one wishes to commune. Attaining the mystery of the mind involves yogic contemplation of and absorption in the Buddha Mahāvairocana and his attendants.

The aspirant was further helped in his quest to identify his buddha-nature with the Cosmic Buddha by means of two sacred drawings, or mandalas, often placed on the Shingon altar. These mandalas, believed to contain all the power of the cosmos, were drawn in accordance with the teaching of Hui-kuo that the doctrines of the Buddha Mahāvairocana were so deep and abstruse that their meanings could be conveyed only in art. One mandala, called the "Diamond Mandala" (based on the *Tattvasaṃgraha* and known in Japanese as *kongō-kai*), portrayed the Buddha Mahāvairocana sitting upon a white lotus in deep contemplation, surrounded by the buddhas of the four regions. This symbolized Mahāvairocana's indestructible, immutable, or potential aspect. The second mandala, called the "Womb Mandala of Great Compassion" (based on the *Mahāvairocana-sūtra* and known in Japanese as *taizō-kai*), revealed Mahāvairocana sitting on a red lotus surrounded by innumerable buddhas, bodhisattvas, and Indian gods, with consorts. This represents the Cosmic Buddha's dynamic manifestation in which he is immanent in everything. Through the correct meditation on these two mandalas it was believed that the aspirant would realize the unity beyond the diversity of the world.

The emphasis of Shingon upon ritual, symbolism, and iconography, coupled with the government's praise of Kūkai and the bestowal upon him of the shrine for the protection of the country, led to a great popularity of Shingon among the Japanese. Many people came to use Shingon rites, believed to control the forces of the cosmos, to ward off evil and bring supernatural help in everyday life. The popularity of Shingon was one of the causes for the growth of Ryōbu Shintō ("Two Aspects Shintō"), which identified Shintō gods with bodhisattvas. While this combination of Esoteric Buddhism with more this-worldly concerns led to some schisms in Shingon, it has continued as one of Japan's strongest Buddhist sects. (Gi.T./F.E.R.)

Popularity of Shingon

Mythology

Myth in Buddhism is used at various intellectual levels in order to give symbolic and sometimes quasi-historical expression to apprehended or presumed religious truths. Accepted on its own terms, Buddhism is a supernatural religion in the sense that, without a buddha to reveal them, the truths remain unknown. Only after human beings have received the Buddha revelation can they proceed apparently by their own efforts. This teaching was explicit in the early schools, in which the revelation was still thought of as historically related to Śākyamuni's mission in the world. Gradually the idea formed, in some schools, of the Buddha's continuous revelation and gracious assistance, deriving from his glorified state of time-transcending Enlightenment. Thus the comparatively simple mythology of the great Buddha myth developed into the far more elaborate mythology of the Mahāyāna.

The acceptance of the mythology, whether early or fully developed, depends upon faith. Without faith the whole religion crumbles to nothing, and nothing is left but a demythologized supposedly historical figure who has no special revelation to give. He becomes a wandering ascetic of ancient India, like the many others known to scholars, and his religion has no explanation. One must, thus, emphasize that it was the extraordinary combination of the historical Śākyamuni and the relevant myth that he was seen to fulfill that set the whole great religious tradition known as Buddhism on its varied historical course.

It has been observed also how myth is continually used at second or even third remove to bolster the primary myth and to give it a more convincing expression. These subsidiary forms of myth include, for example, stories about the recitation of the Buddhist canon soon after Śākyamuni's decease, details of his previous lives, and descriptions of the six spheres of rebirth. Some Buddhist traditions take these subsidiary forms of myth more seriously than others. Within each tradition there are variations among individual adherents. But, even for those Buddhists who are most skeptical, the various myths associated with the Buddha and his saving activity remain central and useful. They rest on premises always provisional but, insofar as they serve the gaining of the chief objective, never really false.

ŚĀKYAMUNI IN LITERATURE AND ART

Traditional literary accounts. The traditional biographies of Śākyamuni, in whatever language they are written, all derive ultimately from early Indian extracanonical rearrangements of the still earlier, scattered canonical accounts of his great acts. The best-known of the Indian "biographies" are the Sanskrit works, the

Mahāvastu, the *Buddhacarita,* and the *Lalitavistara;* the Chinese *Abhiniṣkramaṇa-sūtra,* translated from an Indian original; and the Pāli *Nidānakathā,* as well as the commentary on the *Buddhavaṃsa.* These early works themselves are the result of a continual traditional growth, and to ascertain the dates of their final versions helps in no way to estimate the actual age or reliability of much of the material they contain. All that can be said is that this material agrees substantially with the earliest known fragmentary canonical accounts and that, once presented in coherent biographical form, there are only minor variations in the later "national" versions of the story. The later Sinhalese, Thai, Myanmar (Burmese), and Kampuchean stories are all firmly based on the earlier Pāli versions. The Koreans and Japanese have derived their accounts direct from the Chinese, who in turn derive their traditions, via Central Asia, from Indian sources. The Tibetans, who represent the extreme limits of Indian Buddhist developments, draw their versions from the same earlier Indian versions. The biography of Śākyamuni included by the Tibetan historian Bu-ston (1290–1364) in his *Chos 'byung* (*History of Buddhism*) differs from other traditional accounts only by its listing of the later Mahāyāna doctrines as part of Śākyamuni's teaching program on Earth. All in all, the unity of the mythological and quasi-historical interpretations of the life and death of the "historical" Buddha, in whatever Buddhist country they have been retold, remains impressive.

Myths and biographies of the Buddha

The kernel of truth in the claim of the Theravādin Buddhists of Sri Lanka and mainland Southeast Asia to represent unadulterated "original Buddhism" derives from the fact that they have remained faithful to the early enthusiastic acclamation of Śākyamuni as the one and only Buddha of the present dispensation. Though other buddhas were recognized from a very early date, the attention of the early community was focused almost exclusively on the person and activities of Śākyamuni.

All of the early canonical accounts agree in describing Śākyamuni's experience of Enlightenment as a definitive victory over Māra, the Evil One, and as resulting in a threefold knowledge: that of his own previous births, that of the births and deaths of all other sentient beings, and that of the saving insight that brings final release from the whole unhappy process. However symbolically one may treat the descriptions of the various possible spheres of rebirth among gods, humans, animals, ghosts (pretas), and the denizens of hell, belief in the cosmological myth of continual rebirth is an integral component in the fundamental myth.

Śākyamuni was acclaimed "Great Sage" (Mahāmuni) and Lord (Bhagavat) not because he achieved a state of spiritual equilibrium in the context of ordinary existence but because he attained the supramundane state of nirvana. There are no textual indications that he was ever regarded by his followers as a kind of Socratic sage but rather as a typical perfected yogi (ascetic with magical powers) of his day, possessed—as was then expected—of miraculous powers and divine insight, combined with an altogether extraordinary concern for the spiritual advancement of others. Thus, from the first, his state of Enlightenment, or buddhahood, was recognized as "transcendent" (*lokottara*) and as the transient embodiment of such supramundane knowledge. Śākyamuni was identified with the pre-Buddhist Indian myth of the "Great Man" (Mahāpuruṣa), conceived of as the universal religious teacher who appears on Earth when the circumstances are ripe.

The myth of the Mahāpuruṣa

He was thus accepted as the seventh in an imagined series of previous buddhas. Why the seventh is not known, unless the number was derived from astronomical association, and the question may be pointless from the mythical viewpoint. His contemporary Mahāvīra, leader of the Jains, was linked to a similar series of 24. The essential mythical idea consists not in the numbers but in the notion of a necessary soteriological process. The title Tathāgata, probably meaning "He Who Has Thus Attained," is regularly used by Śākyamuni of himself, and it would seem likely (whatever 19th-century demythologizing scholars might say to the contrary) that he did indeed use this title. Apart from such utter confidence in his achievement, his religious movement would doubtless have died with him.

Not only do buddhas appear at more or less regular intervals, but the final appearance of any buddha is the culmination of a whole series of previous lives, during which he gradually advances toward enlightenment. Such a belief accords with the whole worldview in which Buddhism had its origin, and it may be supposed fairly that Śākyamuni believed this of himself. In any case, the earliest known Buddhist tradition most certainly presented him as so believing. Popular mythology soon set to work to give some tangible substance to the fundamental myth, and no scholar would doubt that the stories of Śākyamuni's previous lives (Jātaka), included in such profusion in the early canonical texts, are accretions, culled from Indian folk literature in order to exploit an opportunity provided by an aspect of the fundamental myth.

Another example of an aspect of the fundamental myth supplemented by later additions concerns Māra, the Evil One, who represented the force of spiritual evil that Śākyamuni was conscious of having confronted and overcome. Māra is explicitly identified as Concupiscence and as Death, the twin foes of all those who strive toward the tranquil and immortal state of nirvana. At the same time, he is identified with various demons and evil spirits, and the texts usually describe him in these terms. It should be noted that the definitive victory over Māra, on whatever spiritual or popular level this may be understood, remains an inalienable element of the myth. It is just as important as the belief, universally attested in the earliest traditions of all Buddhists, in the omniscience and the miraculous powers of Śākyamuni.

Since Śākyamuni's followers were interested in him as a marvelous being and a transcendent Buddha, such historical reminiscences as may have been preserved in the story are incidental to the recounting of such things as the great acts of his previous lives, his miraculous birth in his last life, the drama of his final Enlightenment while sitting under the pipal tree, his stupendous decision to convert and save others (as symbolized by his first sermon in the Deer Park near Vārānasi [Benares]), and his final decease at Kusinārā.

Śākyamuni in art and archaeology. The primary Buddhist monument, both in early Buddhism and in Buddhist usage to the present day, is the stupa, originally a reliquary mound or tumulus. The cult of the stupa may be attested archaeologically from the 3rd century BC onward, but the canonical literary tradition of all Buddhist lands links this all-important cult to the great events associated with Śākyamuni's decease. Mythologically the stupa becomes

The cult of the stupa

Pramod Chandra

Figure 5: Devotees worshiping at a stupa, the monument that symbolizes the Buddha's Parinirvāṇa, or final transcendence. Detail of a Bhārhut stupa railing, mid-2nd century BC. In the Indian Museum, Calcutta.

the supreme symbol of the Buddha in his fully realized state beyond the bonds of mortality. Carved stonework preserved from the 2nd century BC onward, especially from the ancient stupas of Bhārhut and Sāñchi in India, reveals the total identity of the great Buddha myth, as it has been revealed in the texts and the monuments. The scenes portrayed are those not only of the great events of Śākyamuni's last life but also of the great events of his previous births (Jātaka).

It is noteworthy that in the earliest period (viz, the centuries BC) the supramundane lord is represented by symbols—a tree indicating his Enlightenment, a wheel his first preaching, and a miniature stupa his final nirvana. Such was felt to be the sanctity of his being, that, after his birth as Śākyamuni, even before his Enlightenment, he was not physically portrayed. The tree cult itself involved ancient pre-Buddhist traditions that coalesced with the act of the Enlightenment as performed beneath the pipal or bodhi tree. The wheel was both the symbol of the universal monarch (*cakravartin*) and of the Buddha as universal guide and teacher. The stupa cult, with its extraordinary preoccupation with human relics, may have been a special Buddhist development, related to the clearly expressed faith in nirvana as a supramundane state. It is in marked contrast with the usual Hindu (Brahmanic) horror of mortal remains as unclean.

Śākyamuni began to be figured in sculpture in northwestern India from about the 1st century BC onward, and stereotyped presentations of him soon became the model for future use throughout Asia. Common types of Buddha image are those that represent his calling the earth to witness against Māra by touching it with the fingertips of the right hand, the meditating Buddha protected by a cobra's hood, and the Buddha lying on his right side as he enters final nirvana. The Buddha protected by a cobra's hood represents a coalescing of the Buddha myth with the pre-Buddhist cult of snakes as protecting divinities (the naga cult). This coalescence was justified canonically by a legend that recounts an occasion on which the Buddha was protected from a rainstorm by a great naga king named Mucilinda.

Iconographically, the Buddha image was adapted to all of the main scenes of Śākyamuni's life, and while the later stupas in India and Southeast Asia achieve ever greater artistic splendour, they remain fundamentally the symbols of Śākyamuni's transcendence and continue to be decorated by scenes from his previous lives as well as from his last life. Famous examples are Amarāvatī in South India, dating from about the 3rd century AD (some of its stone carvings are preserved in the British Museum), and Borobuḍur in Java, 9th century AD. Borobuḍur is "Mahāyānist" or even "Vajrayānist" in its symbolic structure, but it reveals the close association between later developments and the great Buddha myth of Śākyamuni.

Cave monasteries

Temples and indeed whole monasteries hewn out of the rock were used by Buddhists at least from the 2nd century BC until the 8th century AD and probably later. Early cave monasteries, famous for their temples with internal stupas set in a kind of sanctuary, are Bhājā, Bhedsa, and Kārli, all within reach of Bombay; others famous for the development of the iconography of the Buddha figure are Kanheri (near Bombay), Nāsik, Ellora, and, especially, Ajanta. At Ajanta are also preserved fine murals dating from the 1st century BC to the 9th century AD. These mainly represent Śākyamuni in his last life and in his previous lives as a compassionate bodhisattva.

The traditions of imagery relating to Śākyamuni thrive to this day chiefly in Sri Lanka and the Southeast Asian countries where Theravāda Buddhism prevails, but even in the Mahāyāna countries of Nepal, Tibet, China, Japan, Mongolia, and Korea the same iconographic traditions are preserved whenever an image or painting of Śākyamuni is required. Nowhere, thus, are they really lost so long as Buddhism remains.

CELESTIAL BUDDHAS AND BODHISATTVAS

Literary references. The starting point of all of the later developed Buddhology was the great Buddha myth under one aspect or another. The early idea of a series of buddhas in time, first 7 and later 24, soon allowed for the idea of a future Buddha, the bodhisattva Maitreya. Next was mooted the possibility of buddhas in other buddha lands elsewhere through endless space. Vague popular knowledge in northwestern India of the great Iranian divinity Ahura Mazdā seems to have led to the general belief in a Great Buddha of the West, known as "Infinite Light" (Amitābha) or "Infinite Life" (Amitāyus). From its beginnings in northwestern India this cult spread across Central Asia to China, Korea, and Japan, where it still has an enormous influence.

Five Celestial Buddhas

In Indian beliefs this Western Buddha was balanced with the Buddha of the East, the "Imperturbable" (Akṣobhya), who iconographically is identical with Śākyamuni in the "earth-witness" posture. The cult of the "Imperturbable" Buddha probably derives from the actual Buddha cult at Bodh Gayā, the historical place of enlightenment. In addition to Amitābha of the West and Akṣobhya of the East, there are three others—Vairocana, Ratnasambhava, and Amoghasiddhi—that make up the Five Celestial Buddhas (one for the centre of the universe and one for the other two cardinal directions). Vairocana, the "Illuminator" (centre), is the universal sage or *cakravartin* buddha, indicated by his gesture of preaching and by the symbol of the wheel. Ratnasambhava, the "Jewel Born" (south), represents the Buddha's selfless giving, indicated by the gesture of giving gifts—right hand open, pointing outward and downward. Amoghasiddhi, "Infallible Success" (north), represents the miraculous power to save, indicated by the hand gesture of giving protection—right hand raised, palm outward and pointing upward. These Five Celestial Buddhas, also sometimes called Dhyāni Buddhas, seem—in the early stages of their development—to have been hypostases (concrete manifestations) of various aspects of Śākyamuni.

Celestial bodhisattvas

The cult of Śākyamuni in his previous lives when he was a future buddha (bodhisattva) likewise developed manifold forms. Maitreya, the buddha-yet-to-come, was already known in the earlier period, but from the 1st century onward there was a great cult of celestial bodhisattvas. Bodhisattvas who became especially popular included Mañjughoṣa ("Gentle Voice") or Mañjuśrī ("Glorious Gentle One"), the representative of divine wisdom; Avalokiteśvara, the "Lord of Compassion"; and Vajrapāṇi, "the one who wields the ritual thunderbolt (*vajra*)" and who, as lord of *yakṣas* (a class of local Indian divinities), entered

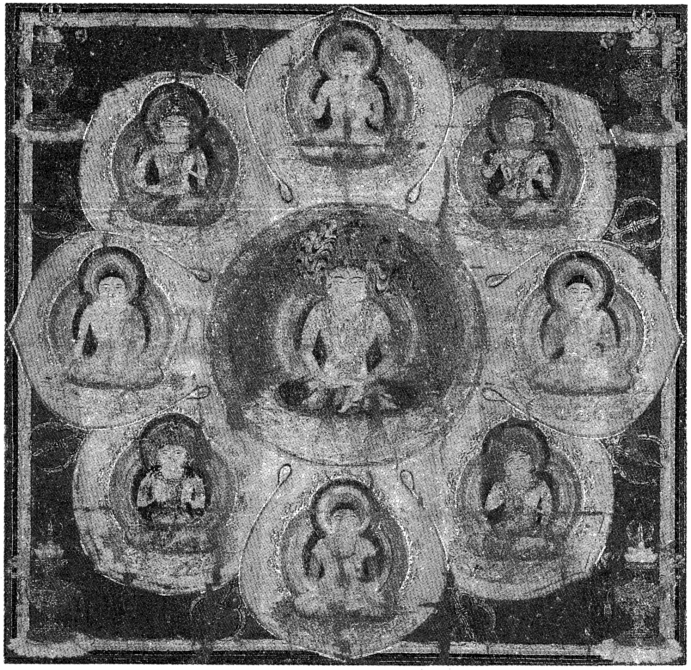
To-ji, Kyoto, Japan

Figure 6: The Five Celestial Buddhas, shown in the centre and at the four points of the compass, with bodhisattvas seated on the petals in between. Central section of Taizōkai, Ryokai Mandara, hanging scroll; early Heian (Jōgan) period, late 9th century AD. In the Tō Temple, Kyōto, Japan.

the pantheon as a great protector. In accord with later developments, these three bodhisattvas were associated with particular buddhas. With the development of the set of Five Buddhas, attendant bodhisattvas were allocated symmetrically to each. Such a balanced systematization was, however, a gradual process and can be traced in the main Mahāyāna sutras.

The *Saddharmapuṇḍarīka* ("Lotus of the Good Law"), a text existing before the 2nd century AD, reveals a great theophany (divine manifestation) of Śākyamuni as glorified lord of the universe. The text of the *Sukhāvatīvyūha* recounts the wonders of Amitābha's Western "Land of Bliss" (Sukhāvatī). The *Karunāpuṇḍarīka* ("The White Lotus of Compassion") is a text concerned with a Buddha Padmottara of the southeast direction but treats also of other buddhas, especially Amitābha and Śākyamuni, as well as of their previous manifestations as bodhisattvas. An important distinction is made between pure Buddha Lands, like that of Amitābha, and impure Buddha Lands, such as the present world in which Śākyamuni appeared. In some texts Śākyamuni is praised as the more noble because out of his great compassion he chose an impure Land.

A new mythology of great importance in East Asian Buddhism developed in association with the pure Buddha Lands. In popular aspiration these replaced for Buddhists the paradises of the ordinary Indian gods, which already formed part of the fivefold or sixfold "wheel of life"—a metaphor for and diagrammatic schematization of the cycle of rebirths. In the case of the pure Buddha Lands, there was the great added advantage of never falling back into unhappy states of existence.

Faith alone suffices to ensure one's rebirth in Amitābha's Western paradise. This particular Buddhist devotion may have begun in northwestern India, whence it passed to Central Asia and on to China and Japan. It had some following in India and Nepal, whence it passed to Tibet, but Amitābha usually remains in later Indian tradition as merely one important member of the Five Buddha group.

The fully developed "Five Buddha" complex found its primary expression in the *Tattvasaṃgraha* ("Symposium of Truth of All the Buddhas"), in which Śākyamuni, as Vairocana, appears as the central buddha. Another related text is the *Mahāvairocana-sūtra,* important for Japanese Esoteric Buddhism (Shingon). The set of Five Buddhas represents the limits of Mahāyāna Buddhological developments, preparing the way for the psychophysical theories of the tantras. The set of five was correlated not only with the centre and four compass points, namely, the macrocosm, conceived as a unity of the Five Great Elements, but also with the microcosm of the human personality understood in terms of the Five Components (skandhas)—*rūpa* (material qualities), *vedanā* (feeling or sensation), *saṃjñā* (perception), samskara (components of consciousness), and *vijñāna* (consciousness)—and with the Five Great Evils (ignorance, wrath, desire, malignity, and envy), typifying normal phenomenal existence. At this stage mythology and psychological symbolization are inextricably bound together.

Correlations of the Five Buddhas

With the tantras, Buddhist mythology began at last to part company with the original Buddha myth and clearly linked up with the Hindu mythology in a scarcely disguised form. Akṣobhya thus has as his fierce Tantric form what is in effect the fierce form of the Hindu god Śiva; in this form he is known under Buddhist names such as Heruka, Hevajra, or Saṃvara. He is known in Japan in this fierce form as Fudō ("Imperturbable"). The Indian god Bhairava, a fierce bull-headed divinity, was adopted by Tantric Buddhists as Vajrabhairava. Also known as Yamāntaka ("Slayer of Death") and identified as the fierce form of the gentle Mañjuśrī, he was accorded quasi-buddha rank.

Some bodhisattvas have become the object of very special devotion. Chief of these is Avalokiteśvara, the Lord of Compassion, who as patron saint of Tibet is believed to reincarnate in the Dalai lamas. As Kuan-yin in China, Kannon in Japan, and Kwanseium in Korea, this bodhisattva coalesces with his feminine counterpart, Tārā, and becomes a kindly madonna. The bodhisattva Kṣitigarbha ("Womb of the Earth"), who was of no great significance in India and hence of none in Nepal and Tibet, attracted a cult as lord of the underworld in Central Asia, whence it spread to China and East Asia generally. Known as Ti-ts'ang in Chinese and Jizō in Japanese, he is Lord of hell and therefore the centre of afterdeath liturgies.

Celestial buddhas and bodhisattvas in art and archaeology. It is mainly from archaeological and artistic remains that scholars have been able to trace the remarkable spread of Mahāyāna Buddhist mythology throughout the whole of Asia from the 1st century AD onward. The main points of departure were northwestern India for Central Asia and East Asia, and the Bay of Bengal, especially the port of Tāmraliptī. Early Mahāyāna developments also affected South India, and thence Sri Lanka and Southeast Asia.

In India itself, Bihar and Bengal remained Buddhist, largely late Mahāyāna and Tantric, until the 13th century, and in Java and Sumatra there is ample iconographic evidence of the popularity of the buddhas, bodhisattvas, and fierce quasi-buddha figures mentioned above. There are even traces in Myanmar, Thailand, and Kampuchea of images and paintings of late Mahāyāna and Tantric divinities. Alone in Southeast Asia, the island of Bali still retains a living but mixed Hindu-Tantric Buddhist culture.

Paintings and figures unearthed during the 20th century in Central Asia (Chinese Turkistan) have revealed the manner in which Buddhist architecture, iconography, and painting passed from northwestern India to China and East Asia. Especially important are the paintings of buddhas and bodhisattvas in the caves of Tun-huang (4th to 10th century AD). Especially popular in China, and hence in Japan and Korea, are Amitābha-Amitāyus, Vairocana, Maitreya, Mañjuśrī, Kṣitigarbha, and Avalokiteśvara (as the goddess Kuan-yin).

The main repository of Indian Mahāyāna and Vajrayāna iconographic traditions is Tibet, where Buddhism was introduced, mainly via Nepal, from the 8th to the 13th centuries. Until the communist takeover of 1959 the Tibetans preserved Indian (Pāla) styles of iconography. They also preserved ancient techniques and styles of Indian Buddhist painting that were modified and enriched in some schools by much later influence from China.

Tibetan art

RECURRENT MYTHIC THEMES

Myths of buddhas and bodhisattvas. The buddhas, the celestial bodhisattvas, and the fierce quasi-buddha manifestations recognized by the followers of the tantras all transcend phenomenal existence in their absolute state while readily involving themselves in it for the sake of suffering living creatures. Although this idea was fully developed only by the Mahāyāna and Esoteric traditions, it has been noted above that the same fundamental conception of a buddha as supramundane, but at the same time operative as world saviour in an immanent sense, belonged to the great Buddha myth constructed around Śākyamuni by the first Buddhist believers. Thus, while a small contemplative elite may always regard any buddha as the impersonal symbol of the ineffable state of enlightenment, most of the faithful have been equally justified in approaching him as a divine saviour. Buddhism has always been a religion of faith, whether of faith in the realizability of final enlightenment or of faith in the buddhas and bodhisattvas as helpers along the way. The Mahāyāna and certainly the Esoteric schools vastly increased the available means of progress to suit all local tastes and propensities, but the fundamental psychology of an accepted dual approach remains scarcely changed. The direct, immediate approach aims at comparatively rapid results and risks wholesale denials, both bodily and mental. The slow approach, through devotion to buddhas and bodhisattvas as divine beings and through the practice of morality in the everyday world, posits religious realities that some members of the spiritual elite in theory deny. But in practice they readily accept them, for they know that a denial would be just as relative to the desired end as the actual acceptance.

Mythic figures in the Three Worlds cosmology. In the early Buddhist tradition Gotama is represented as denying all importance to the questions of whether or not the universe is infinite and whether or not it is eternal. It was enough to realize that normal existence consists of a process of continual birth, death, and rebirth, a process from which, by following the path the Buddha had discovered, one might procure release. If the early texts are correct, however, such an ordinance did not prevent the Buddha, and certainly did not prevent his followers, from accepting the general cosmological beliefs of the time, modified by conclusions drawn from the Buddha's own moral and religious insights.

The three realms

The cosmology, as it was systematized by later Buddhists, included three different realms, all of which were within the confines of samsara (the ongoing cycle of birth, death, and rebirth) and were regulated more or less strictly by the law of karma, according to which good and pious deeds are rewarded while evil and impious deeds are punished. At the top of this universe is the *arūpa-dhātu* ("realm of formlessness"), which has no material qualities. This realm is inhabited by extremely long-lived brahma deities who are absorbed in the deepest levels of yogic trance. Although the existence of the *arūpa* realm is theoretically important for the cosmological system as a whole, the inhabitant deities (when they are distinguished from other brahma deities) play no active role in Buddhist mythology.

Situated just below the *arūpa-dhātu* is the *rūpa-dhātu,* "the realm of form," which has only a remnant of material qualities. This realm is inhabited by brahma deities who are associated with somewhat less exalted levels of yogic attainment than the deities of the *arūpa-dhātu.* Unlike the deities that inhabit the *arūpa* realm, the brahma deities of the *rūpa-dhātu* have a role in Buddhist mythology, particularly in the Buddhist cosmogony through which the lower strata of the cosmological structure are brought back into being after each of the eschatological cataclysms that periodically destroy them. According to one influential version of the primary creation myth, certain brahma deities whose abode is above the destruction begin—as the waters that are left from the old cataclysm start to coagulate below them—to savour the taste of the matter that constitutes these lower strata. As the strata take form, these brahma deities gradually descend into the lower realms and eventually become the first inhabitants of the new earth from whom all the other human inhabitants then descend.

Below the two realms inhabited by brahma deities is the *kāma-dhātu,* "the realm of desire." This realm includes a set of six *gati*s, or destinies, that have played an important role in virtually all Buddhist traditions in Asia. The highest of these six destinies is that of the *devatā* (though both gods and goddesses are included among the *devatā,* the goddesses generally have a secondary role). Within this destiny there are many heavens, each inhabited by many deities. Mythologically the most important are the Tuṣita Heaven where the future Buddha, Metteyya (Sanskrit: Maitreya), awaits the time for his coming to Earth; the Heaven of the Thirty-Three Gods, which is presided over by Inda (Sanskrit: Indra; a deity sometimes called Sakka [Sanskrit: Śakra], who plays a significant mythological role, especially but not exclusively in the Theravāda tradition); and the Heaven of the Four Guardian Kings, who are important protective deities in many Buddhist contexts.

The second of the six *gati*s is the destiny that is enjoyed and suffered by human beings. It is the locus for a myriad of mythic stories about pious monks, nuns, kings, and other laypersons. The third *gati* is the destiny of the asuras; it is relatively uninteresting from a mythic point of view, and it is often omitted. The fourth *gati* is the destiny of the animals. It provides the setting for the description of many fabulous creatures, including nagas, Garuḍa, lions, and elephants.

Pretas and hell beings

The two remaining *gati*s, those of the pretas (hungry ghosts) and the hell beings, are mythically important in two respects. The descriptions provided of the punishments that are inflicted and suffered in these realms are very vivid indeed. In addition, there are widely distributed and well-known mythic stories of compassionate bodhisattvas and Buddhist saints who make journeys to these *gati*s to assuage the torment of those who suffer and to secure their release.

In different areas of Asia new gods, goddesses, and demonic figures were incorporated into the cosmology (for example, in Southeast Asia the great Hindu gods Vishnu [Viṣṇu] and Śiva were often depicted as devas). Despite the new mythic contents that were added, however, the structure itself remained remarkably intact.

Local gods and demons. While the contemplative elite may deny the real existence of gods and demons together with the rest of phenomenal existence, the majority of Buddhists from the earliest times in India, and in other countries where Buddhism has spread, have never neglected indigenous religious beliefs. It has already been noted how Māra, the manifestation of spiritual evil, was presented in the earliest literature in the terms of local demonological beliefs. It is also to be noted that the early stupas and entrances to cave temples were decorated with local male and female deities (usually referred to as *yakṣa* and *yakṣinī*) who were presumably already conceived as converted defenders of the new faith. This proved the easier way of justifying the continuance of the cult of local deities, and it has been employed in varying degrees in every Buddhist land. There thus began to develop a pantheon of minor deities, which from a common original stock has continued to take in new members wherever Buddhism has become an established religion.

The Mahāyāna and Vajrayāna traditions have given these local deities a more ready welcome, even admitting some of their cults as a subsidiary part of the liturgies in honour of buddhas and bodhisattvas. Such favoured deities include Mahākāla, the great black divinity; the mother goddess Hārītī; Kuvera, the god of wealth; and especially Hayagrīva, a fierce horse-faced god who is powerful in driving off unconverted demonic forces.

Throughout the Mahāyāna and Vajrayāna world local deities have become manifestations of various buddhas and bodhisattvas. Perhaps the locus classicus for this process is Japan, where the identification of buddhas and bodhisattvas with indigenous kami has extended from the elite level (for example, in the identification of the buddha Mahāvairocana with the great ancestral Sun goddess, Amaterasu) to temples throughout the country, where a particular buddha or bodhisattva is identified with the kami of the local territory.

Subjugation of Bon deities

In other cases that are equally widespread, local gods and demons have been conquered, converted and brought into the pantheon, or relegated to the periphery (where they may still require propitiation). Perhaps the most interesting example is found in Tibet, where it is commonly believed that Buddhism became established in the 8th century only as the result of the wholesale subjugation of opposing local deities—a subjugation that must, from time to time, be repeated through the performance of rituals marked by their dynamism and ferocity.

Quite as much in Theravāda as in Mahāyāna countries, Buddhism has had to come to terms with local beliefs. In some cases well-organized pantheons have been constructed. For example, in Sri Lanka various local, Hindu, and Buddhist deities hold places within a hierarchy headed by the Buddha himself. In Myanmar the traditional hierarchy of local nats is headed by Thagya Min nat, who is identified with Inda (also known as Sakka). As Inda, he becomes a divine protector of Buddhism, who—in the classical Buddhist cosmology—reigns in the Heaven of the Thirty-Three Gods.

These neatly organized systems, even where they exist, are, however, only a small part of the story. Throughout the various Theravāda countries a wide variety of deities and spirits have been incorporated into the Buddhist world as the inhabitants of particular realms within the Buddhist cosmos or as the guardians of various images, stupas, and temples. At the same time, there are others who, like the demons of Tibet, remain only partially encompassed within the Buddhist domain.

Female deities. In many Buddhist traditions female deities and spirits have been relegated to minor and secondary positions in the pantheon. Among the Theravādins, for example, it is rare for female deities to play a major role. In Sri Lanka, however, the goddess Pattinī is a major deity.

In the Mahāyāna tradition several female deities became major figures. For example, Supreme Wisdom (*Prajñāpāramitā*) is often personified as the Mother of All Buddhas who is manifest especially in Mahāmāyā, the virgin mother of Śākyamuni. Tārā, the saviouress, is a closely related and much more popular figure who has often been taken to be the female counterpart of the bodhisattva Avalokiteśvara. In China and Japan Avalokiteśvara himself gradually assumed a female form. As Kuan-yin (Japanese: Kannon) he/she became probably the most popular figure in the entire panoply of buddhas and bodhisattvas.

Sexual symbolism

It was, however, in the Vajrayāna and Esoteric traditions that female deities became ubiquitous at the highest levels of the pantheon. From the 7th century onward a riot of female divinities began to find their way into certain circles of Buddhist yogis, where they were actually represented by women partners in a special kind of sexual yoga (physical and mental discipline). The process was gradually interpreted as an internal form of celibate yoga, for, in accordance with Vajrayāna and Esoteric theory, enlightenment was achieved by the union of Wisdom and Method, now conceived of symbolically as female and male. Thus it became possible to present supreme buddhahood as the union of a male and female pair and then to represent every celestial buddha or quasi-buddha by a pair of male and female forms. The actual sexual ritual was certainly performed at one time in India and Nepal, seemingly to a very limited extent in Tibet, and perhaps not at all in China and Japan. Nonetheless this form of Tantric symbolism, with its plethora of female buddhas and quasi-buddhas, became a powerful symbolism that virtually all Vajrayāna and Esoteric Buddhists have simply taken for granted as part of their received tradition.

Kings and yogis. The great Buddha myth is a combination of the ideals of universal kingship and universal religious preeminence. This is clearly expressed in the myth of the prophetic utterance of future greatness by the sage Asita—an astrologer who examined auspicious signs on the infant Gotama—over the child bodhisattva. Also in his previous life as Vessantara (Sanskrit: Viśvāntara), Gotama had already realized the perfection of the extraordinary combination of kingship and all-abandoning asceticism. As crown prince, Vessantara was famous for his vast generosity, and, to the despair of his more practical-minded father, he accepted banishment to the forest, where he attained the ultimate of self-abnegation by giving away his children and his wife and in some accounts even his own eyes. These and all the rest were restored to him miraculously, and, responding to the demands of his countrymen, he returned home to become the best of kings. Similarly, the last life of Gotama, up to the time of his great renunciation, is told entirely as a royal story.

Although the practice of Buddhist religion strictly required withdrawal from the world, or at least renunciation of its pleasures, Buddhist monks were, understandably enough, anxious to win royal support. They always needed benefactors, and what better benefactor than a king. Any suggestion of royal benefaction thus resulted in the revival of the "myth" of the vastly generous monarch. Whenever Mahāyāna tendencies have been at work, the notion of the beneficent king as a bodhisattva has been prominent. The most famous example of the mythologized kings is the Indian emperor Aśoka, who facilitated the spread of Buddhism and concerning whom vast legends have grown up. Among other things, he is credited with having built 84,000 stupas. Surrounding countries all claim to have received Buddhism through his mediacy. On a smaller scale legends embellish the life of King Tissa of Sri Lanka (3rd century BC), who presided over the arrival of Buddhism. In the same context one may mention Prince Shōtoku of Japan (d. AD 621) and Srong-brtsan-sgam-po of Tibet (d. AD 650), noting, however, that the enthusiasm of the first for Buddhism is genuinely historical. This is also true of Tibet's two other great "kings of religion": Khri-srong-lde-btsan (reigned 755–797) and Ral-pa-can, assassinated by enemies of the faith in AD 838.

The great stupa of Borobudur, mentioned above, deliberately represents the self-identification of the ruling monarch of Java with the aspiration toward buddhahood. The king presents himself as the bodhisattva par excellence. At the other side of the Buddhist world, the Tibetans developed the same idea when they identified their reincarnating Dalai Lama as a manifestation of their patron "saint," the bodhisattva Avalokiteśvara. By polite mythical fiction the Manchu emperors of China were regarded as manifestations of the bodhisattva Mañjuśrī. As a result of the more restrained nature of their doctrines, Theravādin countries have provided less support for these interesting developments. There such kings as Dhammaceti of Pegu (d. 1491) or Mindon of Myanmar (d. 1878), and even great religious ascetics, remain transient mortal beings of flesh and blood.

The king as bodhisattva

Under the aspect of the pre-Buddhist Indian myth of the ideal perfected yogi, possessed of miraculous powers, however, the greatness of the Buddhist ascetic is a theme well suited to Buddhism everywhere. The early disciples of Śākyamuni, known as arhats when they achieved perfection, were conceived of as miracle-working yogis, and the early canonical literature presents them in this way. This same ideal was acknowledged in Theravādin Sri Lanka, and the Sinhalese claim their share of arhats. But it was in Tibet, which drew on the more developed Indian myth of the "great yogi" (*mahāsiddha*) of the Tantric period (8th to 12th century AD) that this theme showed its most luxurious development. Especially famous are Padmasambhava (also called Guru Rimpoche), an 8th-century Indian yogi credited with having quelled the evil spirits of Tibet, and the strange figure of Pha-dam-pa Sangs-rgyas (d. 1117), a Brahman of South India who became a Buddhist and visited Tibet and possibly China in the 11th century. Doubtless historical, Pha-dam-pa Sangs-rgyas passed out of history into myth with his fantastic powers and equally fantastic longevity. Better known in Europe is the story of the Tibetan yogi Mi-la-ras-pa (1040–1123).

Early in the history of Chinese Buddhism the same mythical tendencies appeared. Bodhidharma (6th century), the founder of Ch'an (Zen) Buddhism, was also an Indian yogi, appearing in quasi-historical, quasi-mythical guise. Subsequently the ideal of the Buddhist sage, as typified by the arhats, coalesced in Chinese thought with the Taoist immortals. In Japan new mythicized stories developed, some associated with the founders of Japanese sects, such as Kūkai and Shinran, others with popular holy men who were the Buddhist counterparts of indigenous shamans and ascetics. Through the continued generation of such new myths and stories, Buddhism was able to move from culture to culture, taking root in each one along the way.

(D.L.S./F.E.R.)

Popular religious practices

Like other great religious traditions, Buddhism has generated a wide range of popular practices. Among these, there are two simple practices that are deeply rooted in the experience of the earliest Buddhist community and that have remained basic to all Buddhist traditions.

Veneration of the Buddha

The first of these is the practice of venerating the Buddha or other buddhas, bodhisattvas, or saints who manifest the same reality. This sometimes takes the form of showing respect, meditating on the qualities of the Buddha, or giving gifts. These gifts are often given to the relics of the Buddha, to images made to represent him, and to other traces of his presence (such as, for example, places where his footprint can supposedly be seen). After the Buddha's death the first foci for this sort of veneration seem to have been his relics and the stupas that were built to hold them. About the beginning of the Common era, anthropomorphic images of the Buddha were produced, and they took their place alongside relics and stupas as focal points for venerating the Buddha. Still later, in the context of the Mahāyāna and Vajrayāna traditions, the veneration of other buddhas and bodhisattvas came to supplement or replace the veneration of the Buddha Gautama. In the course of Buddhist history, the forms have become diverse, but the practice of honouring and even worshiping the Buddha or Buddha-figure has remained a central component in all Buddhist traditions.

The second basic practice is the reciprocal exchange that

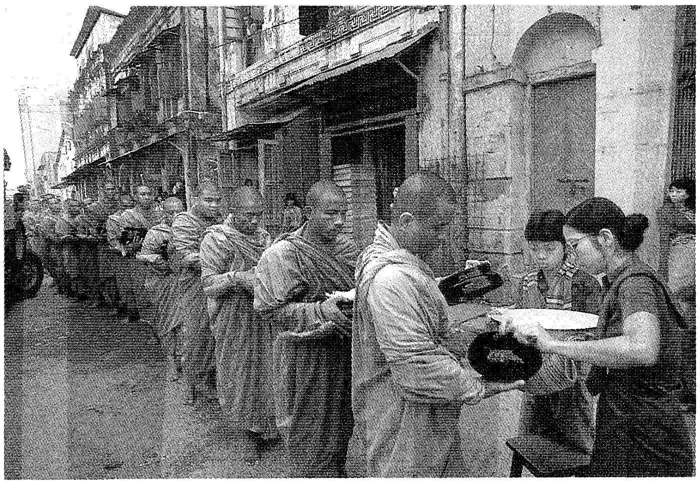

Figure 7: Monks receiving alms from a laywoman in Yangôn, Myanmar.

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takes place between Buddhist monks and Buddhist laypersons. Like the Buddha himself, the monks embody or represent the higher levels of spiritual achievement, which they make available in various ways to the laity. The laity make merit and improve their soteriological condition by giving the monks material gifts that function as sacrificial offerings. Though the exchange is structured differently in different Buddhist traditions, it has remained until recently a component in virtually all forms of Buddhist community life.

Both of these fundamental forms of Buddhist practice appear independently within the tradition. The veneration of the Buddha or Buddha figure is a common ritual often practiced independently of other rituals. So, too, the practice of exchange between monks and laypersons often structures rituals such as the *dāna* (gift-giving) in the Theravāda tradition, which are performed independently of other rituals. Both of these forms of practice, however, are embedded in one way or another in virtually all other Buddhist rituals, including calendric rituals, pilgrimage rituals, rites of passage, and protective rites.

CALENDRIC RITES AND PILGRIMAGE

Uposatha. The four monthly holy days of ancient Buddhism continue to be observed in the Theravāda countries of Southeast Asia. These *uposatha* days—the new moon and full moon days of each lunar month and the eighth day following the new and full moons—have their origin, according to some scholars, in the fast days that preceded the Vedic soma sacrifices. In the Buddhist context laypersons and monks are expected to perform religious duties during the *uposatha* days.

The *uposatha* service typically includes the repetition of the precepts, the offering of flowers to the Buddha image, the recitation of Pāli sutras, meditation practices, and a sermon by one of the monks for the benefit of the visitors. The more pious laymen may vow to observe the eight precepts for the duration of the *uposatha.* These are the five precepts normally observed by all Buddhists—not to kill, steal, lie, take intoxicants, or commit sexual offenses—upgraded to include complete sexual continence, plus injunctions against eating food after noon, attending entertainments or wearing bodily adornments, and sleeping on a luxurious bed. The monks observe the *uposatha* days by listening to the recitation by one of their members of the *Pātimokkha,* or rules of conduct, contained in the *Vinaya Piṭaka* and by confessing any infractions of the rules they have committed.

Commemorations of the Buddha's birth, Enlightenment, and final nirvana

Anniversaries. The three major events of the Buddha's life—his birth, Enlightenment, and entrance into final nirvana—are commemorated in all Buddhist countries but not everywhere on the same day. In the Theravāda countries the three events are all observed together on Vesak, the full moon day of the sixth lunar month (Vesākha), which usually occurs in May. (The Māgha Pūjā takes place three months earlier—on the full moon of February—and celebrates the Buddha's first exposition of the *Pātimokkha.*)

In Japan and other Mahāyāna countries, the three anniversaries of the Buddha are observed on separate days (in some countries the birth date is April 8, the Enlightenment date is December 8, and the death date is February 15). Festival days honouring other buddhas and bodhisattvas of the Mahāyāna and Vajrayāna traditions are also observed, and considerable emphasis is placed on anniversaries connected with the patriarchs of certain sects. Padmasambhava's anniversary, for example, is especially observed by the Rnying-ma-pa sect in Tibet, and the birthday of Nichiren is celebrated by his followers in Japan.

Vassa. The beginning of *vassa,* the three-month rainy-season retreat from July to October, and its conclusion are two of the major festivals of the year among Theravāda Buddhists, particularly in Myanmar, Kampuchea, Thailand, and Laos. The retreat has largely been given up by Mahāyāna Buddhists. It is an accepted practice in countries such as Thailand for a layman to take monastic vows for the *vassa* period and then to return to lay life. Commonly, the number of years a monk has spent in monastic life is expressed by counting up the number of *vassas* he has observed.

The end of *vassa* is marked by joyous celebration. The following month is a major occasion for presenting gifts to monks and acquiring the consequent merit. The *kaṭhina,* or robe-offering ceremony, is a public event during this period and usually involves a collective effort by a village, a group of villages, or a company to bestow gifts on an entire monastery. A public feast and display of

Ewing Galloway

Figure 8: Interior of the Wat Arun (Temple of the Dawn), Bangkok, Thai., during Vesak, the festival commemorating the birth, Enlightenment, and death of the Buddha.

the robes and other presents on a "wishing-tree" are the usual components of the ceremony. The *kaṭhina* season is climaxed by the making and presentation of the *mahākaṭhina* ("great robe"), a particularly meritorious gift that requires the cooperation of a number of people who, theoretically at least, must produce it—from spinning the thread to stitching the cloth—in a single day and night. The robe commemorates the act of the Buddha's mother, who on hearing that he was about to renounce worldly life, wove his first mendicant robes in one night.

All Souls festival. The importance of the virtues of filial piety and the reverence of ancestors in China and Japan have established Ullambana, or All Souls Day, as one of the major Buddhist festivals in those countries. In China, worshipers in Buddhist temples make "boats of the law" (*fa-ch'uan*) out of paper, some very large, which are then burned in the evening. The purpose of the celebration is twofold: to remember the dead and to free and let ascend to heaven the pretas. The pretas are the spirits of those who died as a result of an accident or a drowning and as a consequence were never buried; their presence among men is thought to be dangerous. Under the guidance of Buddhist temples, societies (*hui, Yu-lan-hui*) are formed to carry out ceremonies for the pretas—lanterns are lit, monks are invited to recite sacred verses, and offerings of fruit are presented. An 8th-century Indian monk, Amoghavajra, is said to have introduced the ceremony into China, from where it was transmitted to Japan. During the Japanese festival of Bon, two altars are constructed, one to make offerings to the spirits of dead ancestors and the other to the souls of those dead who have no peace. Odorinembutsu (the chanting of invocations accompanied by dancing and singing) and invocations to Amida are features of the Bon celebrations.

The Japanese Bon festival

New Year's and harvest festivals. New Year's festivals and harvest festivals are examples of Buddhism's involvement in preexisting local traditions. On the occasion of the New Year, images of the Buddha are in some countries taken in procession through the streets. Worshipers visit Buddhist sanctuaries and circumambulate the stupa, or sacred image, and monks are fed and presented with gifts. One of the most remarkable examples of the absorption of local custom was the Smonlam festival in Tibet, celebrated on a large scale in Lhasa until the beginning of Chinese communist rule in the 1950s. The festival was instituted in 1408 by Tsong-kha-pa, the founder of the predominant Dge-lugs-pa sect, who transformed an old custom into a Buddhist festivity. Smonlam took place at the beginning of the winter thaw, when caravans began to set out once again and the hunting season was resumed. The observances included exorcistic ceremonies, performed privately within each family to remove evil forces lying in wait for individuals as well as for the community as a whole, and propitiatory rites, performed to ward off evil such as droughts, epidemics, or hail, during the coming year. During the more public propitiatory rites, the sangha cooperated with the laity by invoking the merciful forces that watch over good order, and processions, fireworks, and various amusements created an atmosphere of hopefulness. Through the collaboration of the monastic community and the laity, a general reserve of good karma was accumulated to see everyone through the dangerous moment of passage from the old year to the new.

The harvest festival celebrated in the Tibetan villages during the eighth lunar month was quite different in nature from the New Year ceremonies. Most commonly, offerings of thanks were made to local deities in rites that were only externally Buddhist. The same interplay between Buddhism and folk tradition is observable elsewhere. In Sri Lanka at harvest time, for example, there is a "first fruits" ceremony that entails offering the Buddha a large bowl of milk and rice.

An integral part of the harvest celebrations in many Buddhist countries is the sacred performance of an episode in the life of a buddha or a bodhisattva. In Tibet, troupes of actors specialize in performances of Buddhist legends. In Thailand, the recitation of the story of Phra Wes (Pāli: Vessantara) constitutes one of the most important festival events of the agricultural calendar.

Buddhist pilgrimage. Within the first two centuries following the Buddha's death, pilgrimage had already become an important component in the life of the Buddhist community. During these early centuries of Buddhist history there were at least four major pilgrimage centres—the place of the Buddha's birth at Lumbinī, the place of his Enlightenment at Bodh Gayā, the Deer Park in Vārānasi (Benares) where he supposedly preached his first sermon, and the village of Kusinārā, which was recognized as the place of his Parinirvāna. During this period, the place of the Buddha's Enlightenment at Bodh Gayā was the most important pilgrimage centre; it continued to hold a preeminent position through much of Buddhist history, and it is the major pilgrimage site of world Buddhism today.

Bodh Gayā

In addition to these four primary sites, major pilgrimage centres have emerged in every region or country where Buddhism has been established. Many local temples have their own festivals associated with a relic enshrined there or an event in the life of a sacred figure. Some of these, such as the display of the tooth relic at Kandy, Sri Lanka, are occasions for great celebrations attracting many pilgrims. In many Buddhist countries famous mountains have become sacred sites that draw pilgrims from both near and far. In China, for example, four such mountain sites are especially important: O-mei, Wu t'ai, P'u-t'o, and Chiu-hua. Each is devoted to a different bodhisattva whose temples and monasteries are located on the mountainside. In many Buddhist regions there are pilgrimages that include stops at a whole series of sacred places. One of the most interesting of these is the Shikoku pilgrimage in Japan, which involves visits to 88 temples located along a route that extends for more than 700 miles.

Buddhist pilgrimages, like pilgrimages in other religions, are undertaken for a wide range of reasons. For some Buddhists, pilgrimage is a discipline that fosters spiritual development; for others it is the fulfillment of a vow—as, for example, after recovery from an illness; and for others it is simply an occasion for travel and enjoyment. Whatever the motivations that support it may be, pilgrimage is one of the most important Buddhist practices.

RITES OF PASSAGE AND PROTECTIVE RITES

Initiation. *Ordination.* Admission to the sangha involves two distinct acts: *pabbajjā,* which consists of renunciation of secular life and acceptance of monasticism as a novice, and *upasampadā,* the official consecration as a monk. The evolution of the procedure is not entirely clear; in early times, the two acts probably occurred at the same time. Subsequently, the *Vinaya* established that *upasampadā,* or full acceptance into the monastic community, should not occur before the age of 20, which, if the *pabbajjā* ceremony took place as early as the age of 8, would mean after 12 years of training. Ordination could not occur without the permission of the aspirant's parents. The initial Pāli formula was "Ehi bhikkhu," "Come, O monk!"

The rite established in ancient Buddhism remains essentially the same in the Theravāda tradition. To be accepted, the postulant shaves his hair and beard and dons the yellow robes of the monk. He bows to the abbot or senior monk, to whom he makes his petition for admittance, and then seats himself with legs crossed and hands folded, pronouncing three times the formula of the Triple Refuge ("I take refuge in the Buddha, I take refuge in the *dhamma,* I take refuge in the sangha.") He repeats after the officiating monk the Ten Precepts and vows to observe them. Thereafter, in the presence of at least 10 monks (fewer in some cases) the postulant is questioned in detail by the abbot—as to the name of the master under whom he studied, whether he is free of faults and defects that would prevent his admission, and whether he has committed any infamous sins, is diseased, mutilated, or in debt. The abbot, when satisfied, thrice proposes acceptance of the petition; the chapter's silence signifies consent. Nuns were once ordained in basically the same way.

The *upasampadā* rite

Bodhisattva vows. In Mahāyāna Buddhism, new rituals were added onto the ceremony of ordination prescribed by the Pāli *Vinaya.* The declaration of the Triple Refuge is as central an assertion as ever, but special emphasis is placed on the candidate's intention to achieve enlightenment and his undertaking of the vow to become a bodhisattva. Five monks are required for the ordination: the head monk, one who guards the ceremony, a master of secrets (the esoteric teachings, such as mantras), and two assisting officiants.

Abhiṣeka. The esoteric content of Vajrayāna tradition requires a more complex ceremony of consecration. In addition to the other ordination rites, preparatory study, and training in yoga, the Tantric neophyte receives *abhiṣeka* (literally, "sprinkling," of water). This initiation takes several forms, each of which has its own corresponding "wisdom" (*vidyā*), rituals, and esoteric formulas; each form of initiation is associated with one of the Five Buddhas of the supreme pentad. The initiate meditates on the *vajra* ("thunderbolt") as a symbol of Vajrasattva Buddha (the Adamantine Being), on the bell as a symbol of the void, and on the mudra (ritual gesture) as "seal." The intent of the initiation ceremony is to produce an experience that anticipates the moment of death. The candidate emerges reborn as a new being, a state marked by his receipt of a new name.

Funeral rites. The origin of the Buddhist funeral observances can be traced back to Indian customs. The cremation of the body of the Buddha and the subsequent distribution of his ashes are told in the *Mahāparinibbāna Sutta* ("Sutra on the Great Final Deliverance"). Early Chinese travelers such as Fa-hsien described cremations of venerable monks. After cremation, the ashes and bones of the monk were collected and a stupa built over them. That this custom was widely observed is evident from the large number of stupas found near monasteries.

With less pomp, cremation is also used for ordinary monks and laymen, though not universally. In Sri Lanka, for example, burial is also common. In Tibet also, because of the scarcity of wood, cremation is rare. The bodies of great lamas, such as the Dalai and Panchen lamas, are placed in rich stupas in attitudes of meditation, while lay corpses are exposed in remote places to be devoured by vultures and wild animals.

Beliefs about the state and journey of the dead

Buddhists generally agree that the thoughts held by a person at the moment of death are of essential significance. For this reason, sacred texts are sometimes read to the dying person to prepare the mind for the moment of death; similarly, sacred texts may be read to the newly dead, since the conscious principle is thought to remain in the body for about three days following death. In Tibetan, Mongolian, and Chinese lamaseries, a lama sometimes recites the famous *Bardo Thödrol* (commonly referred to in English as *The Tibetan Book of the Dead*).

Protective rites. From a very early period in its development, Buddhism has included within its repertoire of religious practices specific rituals that are intended to protect against various kinds of danger and to exorcise evil influences. In the Theravāda tradition protective, exorcistic rituals are closely associated with texts called *paritta*s, many of which are attributed directly to the Buddha. In Sri Lanka and the Theravāda countries of Southeast Asia, *paritta*s are traditionally chanted during large public rituals designed to avert collective, public danger. They are also very widely used in private rituals intended to protect the sponsor against illness and various other misfortunes.

In the Mahāyāna and Esoteric traditions, the role taken by protective and exorcistic rituals was even greater. For example, dharanis (short statements of doctrine that supposedly encapsulate its power) and mantras (a further reduction of the dharani, often to a single word) were widely used for this purpose. Protective and exorcistic rituals that used such dharanis and mantras were extremely important in the process through which the populations of Tibet and East Asia were converted to Buddhism. They have remained an integral part of the Buddhist traditions in these areas, reaching what was perhaps their fullest development in Tibet. (Gi.T./F.E.R.)

Buddhism in the contemporary world

MODERN TRENDS

During the 19th and 20th centuries Buddhism has been forced to respond to new challenges and opportunities that cut across the regional religious and cultural patterns that characterized the Buddhist world in the premodern period. A number of Buddhist countries were subjected to Western rule, and even those that were not felt the heavy pressure of Western religious, political, economic, and cultural influence. Modern rationalistic and scientific modes of thinking, modern notions of liberal democracy and socialism, and modern patterns of economic organization were introduced and became important elements in the thought and life of Buddhists and non-Buddhists in these countries. In this situation the Buddhists' response was twofold. They came to associate Buddhism with the religious and cultural identity that they sought to preserve and reassert in the face of Western domination. In addition, they sought to initiate reforms that would make Buddhism a more appealing and effective force in the modern world.

The Buddhist concern to challenge Western domination manifested itself both in the specifically religious and in the religiopolitical sphere. In the former, Buddhists used a variety of measures to meet the challenge posed by the presence of Western Christian missionaries, often adopting modern Christian practices such as the establishment of Sunday schools, the distribution of tracts, and the like. They also attempted to strengthen the Buddhist cause through the initiation of Buddhist missions, including missions to the West, and through ecumenical cooperation among various Buddhist groups. Organizations such as the World Fellowship of Buddhists (founded 1950) and the World Buddhist Sangha Council (1966) were established to promote cooperation among Buddhists from all countries and denominations.

International Buddhist organizations

In the religiopolitical sphere, many Buddhist leaders—including many politically active monks—sought to associate Buddhism with various nationalist movements that were struggling to achieve political, economic, and cultural independence. Where these leaders and the nationalist causes with which they associated themselves have been successful (as, for example, in Thailand), Buddhism has retained a central role in political life. Where they were superseded by other forces (as in China), Buddhism has been relegated to the periphery.

Three emphases have been especially important in the various reform movements. First, many Buddhist leaders have put forward a highly rationalized, Protestant-type interpretation of Buddhism that deemphasizes the supernormal and ritualized aspects of the tradition and focuses on the supposed continuity between Buddhism and modern science and on the centrality of ethics and morality. This interpretation, according to its proponents, represents a recovery of the true Buddhism of the Buddha.

A second, closely related emphasis that has been prominent among modern Buddhist reformers represents Buddhism as a form of religious teaching and practice that provides a basis for social, political, and economic life in the modern world. In some cases the focus has been on Buddhist ideas that supposedly provide a religious grounding for an international order supporting world peace. Other reformers have presented Buddhism as a basis for a modern democratic order or have advocated a Buddhist form of socialism.

Finally, Buddhist reformers have initiated and supported movements that give the Buddhist laity (and in some cases Buddhist women) a much stronger role than they have had in the past. In the Theravāda world, lay societies have been formed and lay-oriented meditation movements have enjoyed great success. In East Asia an anticlerical, lay-oriented trend that was evident even before the modern period has culminated in the formation and rapid expansion of new, thoroughly laicized Buddhist movements, particularly in Japan.

CHALLENGES AND OPPORTUNITIES

The status of contemporary Buddhist communities and the kinds of challenges those communities face differ radically from area to area. Five different kinds of situations can be identified.

First, there are a number of countries where previously well-established Buddhist communities have suffered se-

Buddhism under communist governments

vere setbacks that have curtailed their influence and seriously sapped their vitality. This kind of situation prevails primarily in countries ruled by communist governments where Buddhism has, for many decades, been subjected to intense pressures that have undercut its institutional power and weakened its influence on large segments of the population. This has happened in the Mongol areas of Central Asia, in China (outside of Tibet), in North Korea, and, to a lesser extent, in Vietnam.

Second, there are places where well-established Buddhist communities have suffered similar setbacks but have retained the loyalty of large segments of the population. Perhaps the most vivid example is Tibet, where the Chinese communists have implemented anti-Buddhist policies that, despite their brutality, have failed to break the bond between Buddhism and the Tibetan sense of identity. In Kampuchea and Laos, similarly, communist rule (including even the reign of terror imposed by the Pol Pot regime that controlled Kampuchea from 1975 to 1979) does not seem to have broken the people's loyalty to Buddhism.

Third, there are situations in which the Buddhist community has retained a more or less accepted position as the leading religious force and has continued to exert a strong influence on political, economic, and social life. This is the case in Sri Lanka and Myanmar, where Buddhism is the dominant religion among the Sinhalese and Burman majorities, and in Thailand, where more than 90 percent of the population is counted as Buddhist. In Sri Lanka and Myanmar, ethnic conflict and (especially in Myanmar) authoritarian rule and economic stagnation have resulted in political instability that has had a disruptive effect on the local Buddhist communities. In Thailand, however, Buddhism has a firm position within a relatively stable and rapidly modernizing society.

The fourth type of situation is one in which well-developed Buddhist traditions are operating with a considerable degree of freedom and effectiveness in societies where Buddhism plays a more circumscribed role. This situation prevails in several of the Pacific Rim countries, including South Korea, Taiwan, Hong Kong and Singapore, and to a lesser extent in Southeast Asian countries such as Malaysia and Indonesia, where Buddhism is practiced by significant numbers of overseas Chinese. The primary example, however, is Japan, where Buddhism has continued to play an important role. In the highly modernized society that has developed in Japan, many deeply rooted Buddhist traditions, such as Shingon, Tendai, the Pure Land schools, and Zen, have persisted and have been adapted to changing conditions. At the same time, new Buddhist sects such as Rissho-Kosei-Kai and Soka-gakkai have gained millions of converts in Japan and throughout the world.

Finally, new Buddhist communities have developed in areas where Buddhism disappeared long ago or never existed at all. Thus in India, where Buddhism had been virtually extinct since at least the 15th century, new Buddhist societies have been formed by Indian intellectuals, new Buddhist settlements have been established by Tibetan refugees, and a significant Buddhist community has been founded by converts from the so-called scheduled castes. In the West (particularly but not exclusively in the United States), important Buddhist communities have been established by immigrants from East and Southeast Asia. Buddhist influences have penetrated into many aspects of Western culture, and communities of Buddhist converts are active.

For more than two millennia Buddhism has been a powerful religious, political, and social force, first in India, its original homeland, and then in many other lands. It remains a powerful religious, political, and cultural force in many parts of the world today. (F.E.R.)

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The classic study of the early Buddhist schools is ANDRÉ BAREAU, *Les Sectes bouddhiques du petit véhicule* (1955). The doctrine of the Sarvastivādins, which is especially important for later Buddhist philosophy, is presented by TH. STCHERBATSKY, *The Central Conception of Buddhism and the Meaning of the Word "Dharma"* (1923, reissued 1974). The early Theravāda teaching is discussed in JOHN ROSS CARTER, *Dhamma: Western Academic and Sinhalese Buddhist Interpretation* (1978); and in the excellent work by STEVEN COLLINS, *Selfless Persons: Imagery and Thought in Theravāda Buddhism* (1982). Theravāda ethics are discussed in WINSTON L. KING, *In the Hope of Nibbana* (1964). King has also published *Theravāda Meditation: The Buddhist Transformation of Yoga* (1980), a significant work. The teachings of an important contemporary reformer are comprehensively presented by LOUIS GABAUDE, *Une Herméneutique Bouddhique contemporaine de Thaïlande: Buddhadasa Bhikku* (1988).

EDWARD CONZE, *Buddhist Thought in India: Three Phases of Buddhist Philosophy* (1962, reprinted 1983), is a good introduction to the development of Buddhist philosophy in India. More detailed studies include the classic L. DE LA VALLÉE POUSSIN, *Le Dogme et la philosophie du bouddhisme* (1930); and D. SEYFORT RUEGG, *The Study of Indian and Tibetan Thought* (1967), a difficult but important work. The rise of Mahāyāna is considered by PAUL WILLIAMS, *Mahāyāna Buddhism: The Doctrinal Foundation* (1989). Among the several excellent books on the Buddhist philosopher Nāgārjuna, two deserve special mention: T.R.V. MURTI, *The Central Philosophy of Buddhism* (1955, reissued 1980); and FREDERICK I. STRENG, *Emptiness: A Study in Religious Meaning* (1967). A central topic is treated in WILLIAM MONTGOMERY MCGOVERN, *A Manual of Buddhist Philosophy,* vol. 1, *Cosmology* (1923, reprinted 1977); and RANDY KLOETZLI, *Buddhist Cosmology: From Single World System to Pure Land: Science and Theology in the Images of Motion and Light* (1983). For studies of the Indo-Tibetan Tantric tradition, see SHASHI BHUSHAN DASGUPTA, *An Introduction to Tantric Buddhism,* 2nd ed. (1958, reprinted 1974); and ALEX WAYMAN, *The Buddhist Tantras: Light on Indo-Tibetan Esoterism* (1973), a collection of more specialized essays.

Among the East Asian traditions, Ch'an (Zen) has been the most adequately studied. See, for example, HEINRICH DUMOULIN, *Zen Buddhism: A History,* 2 vol. (1988–89; originally published in German, 1959); T.P. KASULIS, *Zen Action/Zen Person* (1981), a lucid presentation; and WINSTON KING, *Death Was His Kōan: The Samurai-Zen of Suzuki Shōsan* (1986). Significant books that introduce the teachings of other East Asian schools include FRANCIS H. COOK, *Hua-yen Buddhism: The Jewel Net of Indra* (1977); MINORU KIYOTA, *Shingon Buddhism: Theory and Practice* (1978); and ALFRED BLOOM, *Shinran's Gospel of Pure Grace* (1965).

In recent years Buddhists have begun to address Western philosophy and Christian thought in a serious way. KEIJI NISHITANI, *Religion and Nothingness* (1982), is an expression of the Kyōto school of Japanese Buddhist thinkers who have taken Western philosophy very seriously. GUNAPALA DHARMASIRI, *A Buddhist Critique of the Christian Concept of God* (1974, reissued 1988), is a Sri Lankan Buddhist's more polemical engagement with Christian theology.

Sangha: The basic "constitution" of the Buddhist sangha is discussed in JOHN CLIFFORD HOLT, *Discipline, the Canonical Buddhism of the Vinayapiṭaka* (1981). SUKUMAR DUTT, *Early Buddhist Monachism, 600 B.C.–100 B.C.* (1924, reissued 1984), traces the early history of the sangha in India. H.D. SANKALIA, *The University of Nalanda,* 2nd rev. and enlarged ed. (1972), recounts the history of one of the most important of all Buddhist monastic establishments.

The best study of premodern monasticism in the Theravāda world is R.A.L.H. GUNAWARDANA, *Robe and Plough: Monasticism and Economic Interest in Early Medieval Sri Lanka* (1979). Discussions of more recent monastic situations include E. MICHAEL MENDELSON, *Sangha and State in Burma: A Study of Monastic Sectarianism and Leadership* (1975); and JANE BUNNAG, *Buddhist Monk, Buddhist Layman: A Study of Urban Monastic Organization in Central Thailand* (1973). A more idyllic picture of Theravāda monastic life is depicted by MICHAEL CARRITHERS, *The Forest Monks of Sri Lanka: An Anthropological and Historical Study* (1983). The relationship between the monastic community and society is definitively treated in HEINZ BECHERT, *Buddhismus, Staat und Gesellschaft in den Ländern des Theravāda-Buddhismus,* 3 vol. (1966–73).

Good book-length discussions of Buddhist monasteries and monastic life in Central and East Asia are fewer, but the following can be recommended: ROBERT J. MILLER, *Monasteries and Culture Change in Inner Mongolia* (1959); J. PRIP-MØLLER, *Chinese Buddhist Monasteries* (1937, reissued 1967); J.J.M. de GROOT, *La Code du mahâyâna en Chine: son influence sur la vie monacale et sur le monde laïque* (1893, reprinted 1980); and HOLMES WELCH, *The Practice of Chinese Buddhism: 1900–1950* (1967), a detailed discussion of monastic institutions. One aspect of monastic life in Japan has been presented by DAISETZ TEITARO SUZUKI, *The Training of the Zen Buddhist Monk* (1934, reprinted 1974). NEIL MCMULLIN, *Buddhism and the State in Sixteenth-Century Japan* (1984), is a historical study.

Myths and legends: Except for the myth of Śākyamuni, Buddhist myths and legends have not, for the most part, received extended scholarly discussion. One of the few studies of Theravāda materials is TREVOR LING, *Buddhism and the Mythology of Evil: A Study of Theravāda Buddhism* (1962). JOHN S. STRONG, *The Legend of King Aśoka: A Study and Translation of Aśokāvadāna* (1983), though it has a pan-Buddhist relevance, focuses on a Sanskrit Hīnayāna tradition in superb fashion. For Mahāyāna and Vajrayāna/Esoteric mythology, two major sources are available: the *Mythologie asiatique illustrée* (1928), which provides extensive coverage of the relevant material; and ALICE GETTY, *The Gods of Northern Buddhism: Their History, Iconography, and Progressive Evolution Through the Northern Buddhist Countries,* 2nd ed. (1928, reprinted 1988). ALICIA MATSUNAGA, *The Buddhist Philosophy of Assimilation: The Historical Development of the Honji-Suijaku Theory* (1969), discusses the way in which Buddhism assimilated local deities and cults in China and particularly in Japan.

Images and symbols: Two books that provide excellent Pan-Asian introductions are DIETRICH SECKEL, *The Art of Buddhism* (1964; originally published in German, 1962); and DAVID L. SNELLGROVE (ed.), *The Image of the Buddha* (1978). A less readable collection, though on a very important topic, is ANNA LIBERA DALLAPICCOLA and STEPHANIE ZINGEL-AVE LALLEMANT (eds.), *The Stupa: Its Religious, Historical, and Architectural Significance* (1979).

Although there are many studies of Buddhist images and symbols, few of them focus on the distinctively Buddhist significance of such images. Among the exceptions are ANANDA K. COOMARASWAMY, *Elements of Buddhist Iconography,* 3rd ed. (1979); an important book-length article by PAUL MUS, "Le Buddha Paré: son origine indienne: Çākyamuni dans le mahāyānisme moyen," *Bulletin de l'École Française d'Extrême-Orient,* 28:(1–2)153–280 (January–June 1928); MARIE-THÉRÈSE MALLMANN, *Introduction à l'iconographie du tântrisme bouddhique* (1975); a number of the articles contained in LUIS O. GÓMEZ and HIRAM W. WOODWARD, JR. (eds.), *Barabudur: History and Significance of a Buddhist Monument* (1981); FERDINAND DIEDERICH LESSING, *Yung-Ho-Kung: An Iconography of the Lamaist Cathedral in Peking, with Notes on Lamaist Mythology and Cult* (1942), a very complex work; RYUJUN TAJIMA, *Les Deux Grand Mandalas et la doctrine de l'ésotérisme shingon* (1959); and SHIN'ICHI HISAMATSU, *Zen and the Fine Arts* (1971, reissued 1982).

Popular religious practices: Though many books on various topics mention popular religious practices, few full-length studies are available. STEPHEN F. TEISER, *The Ghost Festival in Medieval China* (1988), a reconstruction and analysis, is the best work available on premodern rites and ceremonies. Among the full-length studies based to a considerable extent on personal observation, three of the most interesting focus on Thailand and Laos: KENNETH E. WELLS, *Thai Buddhism: Its Rites and Activities* (1960, reprinted 1982); STANLEY J. TAMBIAH, *Buddhism and the Spirit Cults in North-East Thailand* (1970); and MARCEL ZAGO, *Rites et cérémonies en milieu bouddhiste lao* (1972). There are four quite different but equally interesting works that focus on practices in Tibet and neighbouring regions: ROBERT B. EKVALL, *Religious Observances in Tibet: Patterns and Function* (1964); STEPHEN BEYER, *The Cult of Tārā: Magic and Ritual in Tibet* (1973), an important descriptive and interpretive work; and two anthropological studies by SHERRY B. ORTNER, *Sherpas Through Their Rituals* (1978) and *High Religion: A Cultural and Political History of Sherpa Buddhism* (1989).

Contemporary trends: Since the mid-1960s, two broad surveys of Buddhism in the modern world have been published. ERNST BENZ, *Buddhism or Communism: Which Holds the Future of Asia?* (1965; originally published in German, 1963), though rather superficial and seriously dated, contains interesting material on Buddhist reform movements that have crossed national boundaries. HEINRICH DUMOULIN and JOHN C. MARALDO (eds.), *Buddhism in the Modern World* (1976; originally published in German, 1970), includes essays on Buddhism in various regions and countries in Asia, Europe, and America.

Several books have been published that deal with the new Buddhist communities that have been established in India and in the West: TREVOR LING, *Buddhist Revival in India: Aspects of the Sociology of Buddhism* (1980); KŌSHŌ YAMAMOTO, *Buddhism in Europe: Report of a Journey to the West, in 1966, of an Eastern Buddhist* (1967); EMMA MCCLOY LAYMAN, *Buddhism in America* (1976); CHARLES S. PREBISH, *American Buddhism* (1979); and LOUISE H. HUNTER, *Buddhism in Hawaii: Its Impact on a Yankee Community* (1971).

(F.E.R.)

Buenos Aires

Buenos Aires, the federal capital of Argentina, is one of the most populous cities in the world. The city proper, which is also the Federal District, is located on the western shore of the Río de la Plata and covers an area of 77 square miles (200 square kilometres). The total metropolitan area, however, occupies some 1,500 square miles. According to tradition, 16th-century Spanish sailors named the port for their patron saint, Santa María del Buen Aire (St. Mary of the Good Air).

The city is the national centre of commerce, industry, politics, and culture. Its wealth and influence overshadow the life of the rest of the nation, and it also presents Argentina with its severest economic and social problems. In addition, because of its location, Buenos Aires often has been a centre for political and social unrest. The Argentine poet and philosopher Ezequiel Martínez Estrada has called the city "The Head of Goliath," a metaphor that illustrates its imbalance in relation to the rest of the country, much like a large-headed giant with a feeble body.

Buenos Aires, which celebrated its 400th anniversary in 1980, is beset with many of the same problems encountered by other large cities. Growing industry is causing increased pollution; limited land area in the city proper has led to the construction of high-rise buildings, adding to urban congestion; rapid expansion is hampered by inadequate public services; increased migration from the countryside is creating a growing urban population composed of the poor and unskilled. Adding to the complexity of daily living is a high rate of inflation, which hampers efforts to improve conditions. There are, in spite of all this, signs of progress and burgeoning strength in the city.

This article is divided into the following sections:

Physical and human geography

CHARACTER OF THE CITY

Even the most casual visitor to the city of Buenos Aires is impressed with its energy and bustle. Tourists who stay only a day or two may walk along the Calle Florida, a pedestrian mall lined with elegant shops, and admire the wealth of goods for sale and the number of elegant, well-dressed people. They can enjoy bargain shopping—mainly of world-class leather goods and woolen clothing—and the steak dinners available in numerous restaurants for considerably less than what the same meal might cost in the United States or in Europe. At night the visitor to the city may be taken to a *boite* ("nightclub") for an evening of tango, the emotional dance that originated in the lower-class areas of the city and is said to reflect the essence of the soul of the *porteño,* the name given to a resident of Buenos Aires.

Origin of the tango

A brief tour usually includes La Boca, a picturesque section at the mouth (Spanish *boca*) of the Riachuelo River, where the city's first settlers landed. The section is filled with Italian restaurants, and some streets, such as the Caminito, are lined with wooden houses painted in bright, gay colours. La Boca, now an artists' colony, was the site of the city's first meat salting plants, which brought great wealth to Buenos Aires in the 19th century.

The visitor will also notice the broad avenues that define the limits between districts in the central city and that are lined by seemingly endless rows of apartment houses. During rush hours these avenues are clogged with traffic, a portion of which is created by a peculiar microbus, called the colectivo, an Argentine invention that is one of the city's distinctive features. Half the size of a typical city bus, it is usually seen crammed with people and often barely pauses as passengers jump on and off. The drivers, usually one of the *socios,* or owners, of the cooperative that operates the line, are often colourful characters, noted for their frequent lively commentary on everything from weather to politics and football (soccer). Their aggressiveness on the road is legendary. Depicted as high-strung but competent, the colectivo driver has become a symbol of the frenetic pace of life in the city.

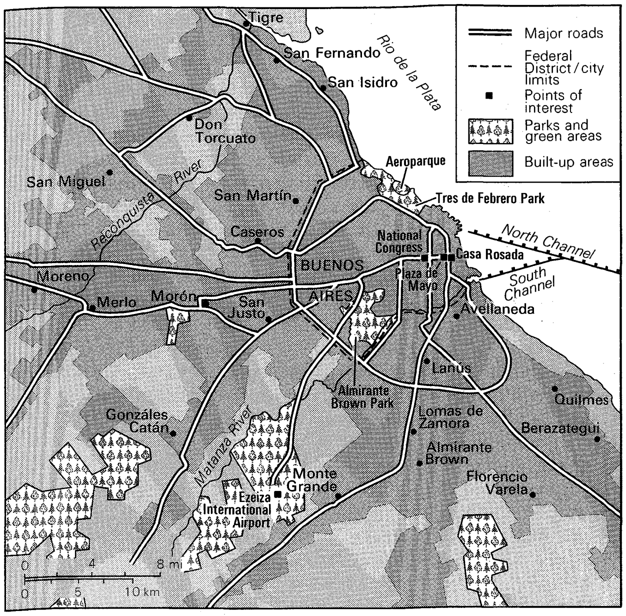
The colectivo

Having observed some elements of the city's essence, the visitor may miss others. Few tourists, for instance, ever get out of the *centro,* the old city centre, which occupies only a fraction of the metropolitan area. Almost everything attractive to the tourist lies there. In addition to the fancy shops, the *centro* is the site of virtually all of the major financial institutions and corporate headquarters. Most of the first-run motion-picture and live theatres are clustered within a four- or five-block stretch on the Avenida Corrientes and the Calle Lavalle, which forms the centre of a crowded entertainment district that is ablaze with light at night.

Buenos Aires cannot be fully known and understood, however, without a visit to the suburban hinterland and the many, varied neighbourhoods that give the city its flavour and explain its relationship to the rest of the country. This helps bring into focus the lives of the people who live in a sprawling city that by the late 20th century was growing nearly five times faster than the country as a whole. Buenos Aires is a city of distinctive neighbourhoods that have their own meeting places, generally coffeehouses or bars. This is a tradition rooted in the colonial period, when the centre of each neighbourhood was a general store and bar known as the pulperia.

THE LANDSCAPE

The city site. The city of Buenos Aires is located at the northeastern edge of the flat plain known as the Pampa, which occupies the heartland of Argentina. It is situated at the point where the Paraná River delta widens to meet the Río de la Plata. The eastern and northern limits of the metropolitan area are defined by the Río de la Plata, and the city's most prominent physical characteristics are the numerous small rivers that flow through its periphery. The



Buenos Aires and surrounding area.

centre of the city lies on a bluff overlooking the Río de la Plata, and to the south flows another river, the Riachuelo, the banks of which mark the other high land in the city. The rest of the city is laid out on the floodplains of the rivers, virtually without significant elevations.

Seasonal temperatures

Climate. The temperate climate of the city is characteristic of the Río de la Plata's coastal plain. The city is hot and humid during the summer months of December to February, with a mean high temperature of 83°F (28°C). The autumn and spring are characterized by fluctuating temperatures and quickly changing weather. The winter months of June to August are mild but humid, with a mean temperature of 52°F (11°C). The average annual temperature is slightly higher than 60° F (16° C). Frosts occur from May to September, but snowfall is extremely rare. Winds are generally of low velocity and are more frequent during the season of electrical storms, between September and March. Rainfall is heaviest in March.

The city layout. The metropolitan area is divided into the Federal District, established in 1880, and the surrounding suburbs, or *partidos.* The Federal District contains less than one-third of the population of the metropolitan area, a proportion that shrinks as the suburbs continue to attract industry and residential communities. The city is divided into districts that coincide mostly with the old neighbourhoods, or barrios. The *centro* is built on the original colonial foundation. It has narrow streets laid out at 90-degree angles to form a grid pattern. This rectilinear pattern holds for more than 20 square blocks, an area that defined the limits of the city until the late 19th century. Since that time, expansion has been less planned, and the pattern of streets is less regular. The limits of the Federal District are marked by the Riachuelo River and the Avenida General Paz, which was opened in 1941 after nearly a decade of construction.

The modern city developed outward from the Plaza de Mayo, a historic square flanked by the Town Hall (Cabildo), which dates from the 18th century, and the Government House, commonly called the Casa Rosada. The Casa Rosada faces west, up the broad Avenida de Mayo, which leads directly to the Plaza del Congreso and the National Congress building, built in the early 20th century. All distances on national highways are measured from the zero-kilometre point in the small square across from the building. The Plaza de Mayo is also the site of the Metropolitan Cathedral, another monumental building from the colonial period. The rest of what remains of colonial Buenos Aires lies mostly south of the plaza, in San Telmo, or Barrio Sur, which, after nearly a century of neglect and decay, has been restored and gentrified.

Architectural styles

Architecturally, the city can be divided into four residential styles. The most common is a structure that began as a single-family dwelling along the street, with an interior patio or garden and rows of small rooms down either side that lead to a kitchen. These houses are attached one to another to form an unbroken facade at the sidewalk. As population density increased in the early 20th century, this basic house was broken up into smaller units and gave rise to a second style, a two- and three-story version known as the *petit hotel,* which was neither as wide nor as deep as its predecessor. The lots on which these houses were constructed defined the size of the first generation of high-rise apartment houses that now dominate the middle-class sections of the city, including Palermo, Recoleta, and Retiro, which, together with other neighbourhoods, are collectively called Barrio Norte. These high rises, representing a third style, were built one next to the other, stretching for block after block in the northern sector of the city. In Belgrano, just north of Barrio Norte, these apartment houses are freestanding; many are as large as city blocks, with their own gardens, because they were built on the lots of single-family detached houses that were common in outer districts of the capital and in the suburban *partidos.*

The *villas miserias*

The fourth residential style is the corrugated metal shack, typical of the shantytowns that have come to constitute about one-fourth of the homes in the metropolitan area and house about 30 percent of the population. These settlements are characterized by their precarious tenure and the absence of basic public services. While such settlements are common to large cities, they have become a significant aspect of the urban landscape in Buenos Aires only since the 1960s. These *villas miserias,* or "popular settlements" as they are called officially, are largely inhabited by rural migrants who felt that they could improve their condition by moving to the city. Because these migrants were utterly without resources, they had little choice but to move onto unoccupied land that was otherwise undesirable.

THE PEOPLE

Buenos Aires is often described as Latin America's most European city. The population is made up largely of the descendants of immigrants from Spain and Italy who came to Argentina in the late 19th or early 20th century. There are also significant minorities of Germans, British, Jews from central and eastern Europe, and Middle East-

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The Casa Rosada on the Plaza de Mayo, Buenos Aires.

ern peoples, who are known collectively as *turcos*. Roman Catholicism is the predominant religion.

Since the 1930s, most of the immigrants to the city have come from the north of Argentina, where the population is predominantly mestizo (mixed Indian and European). Today, the mestizos make up between one-fourth and one-third of the population in the metropolitan area. It is mostly the mestizos who live in the poorest sections of the city, in the *villas miserias* and the distant suburbs. The area's black and mulatto population is negligible in size.

Cultural and linguistic influences

While there are no ethnic neighbourhoods, strictly speaking, many of the smaller minorities tend to settle close to one another in tightly knit communities. Villa Crespo, for example, is known as a Jewish neighbourhood; the Avenida de Mayo is a centre for Spaniards; and Flores is the home of many *turcos*. The assimilation of these groups has been less than complete, but the Argentine identity has been flexible enough to allow ethnocentric mutual-aid societies and social clubs to emerge. Even the dominant Spanish language has been affected by other European cultures and has undergone changes; in the slums and waterfront districts an Italianized dialect has emerged, and Italian cuisine is popular in the city.

THE ECONOMY

The importance of Buenos Aires, the nation's chief port, to the Argentine economy is related to the country's overwhelming dependence on the production and export of agricultural commodities. Buenos Aires is the country's chief point of consumption, processing, and shipping. Unlike much of the rest of the country, the city has a varied economy, which helps it maintain a degree of stability despite the rampant inflation that affects the entire nation.

The port

The port of Buenos Aires is the largest in South America. Ships from all parts of the world stream in to deliver machine goods and consumer durables and to take away grains or agricultural by-products, such as food oils. And yet, as with so much of the city, the port facilities are old and inefficient. At harvest time it is common to see long lines of trucks at the port entrance, waiting to transfer their loads to one of the ships lying offshore, which in turn are waiting to use the narrow channel into the port. Serious proposals to move the port to another, better harbour have met with little response.

Industry. Heavy industry has grown much faster in metropolitan Buenos Aires than elsewhere in Argentina. More than one-third of the industrial capacity of the country is in the Buenos Aires metropolitan area, and about one-half of all manufacturing jobs are located there. During the 1950s there was great emphasis on the production of consumer goods aimed toward making the city self-sufficient in this area. The main industries include food processing (meat, fish, and grain), metalworking, automobile assembly, oil refining, printing and publishing, and the manufacture of textiles, beverages, paper, and chemicals. Service industries account for more than 20 percent of the jobs in the metropolitan area. The area consumes more than half of all the electric and natural gas energy produced in the country. Inside the city proper larger industrial and manufacturing establishments are mostly located in the barrio of Barracas, but they are becoming increasingly more concentrated in the southern suburbs.

Finance and commerce. The city virtually monopolizes the banking activity of the country. Banks in Buenos Aires hold the largest share of the nation's bank deposits, a large portion of which are held by foreign investors. The Banco Central de la República Argentina controls the federal banking system. The Buenos Aires stock exchange, along with specialized markets for meat, cattle, fruit, and grain, makes the city the dominant stock and commodities trading centre in the country. The financial district is concentrated just north of the Plaza de Mayo.

Transportation. Buenos Aires has developed one of the world's better urban transportation systems, although it evolved serendipitously around the unique colectivo. Travelers are seldom more than one block from a bus, and often they have a choice of buses to take. On the other hand, the subway system was designed for the Buenos Aires of the 1940s. Its adequacy in the modern city has been diminishing, given the two socially distinct, yet similar patterns of commuting. Most professionals, businessmen, and other white-collar workers commute to the centre of town by car or train from the northern zones. Blue-collar workers commute across town, from residential to industrial sections, generally by colectivo. Though most city traffic is regulated by automatic traffic lights, the city's residents are notorious for ignoring them. For many years, two major streets, Calle Florida and Calle Lavalle, were traditionally closed to motor traffic during part of the day to allow for a free flow of pedestrians. Calle Florida, however, has been reserved for pedestrians at all times.

Inner-city traffic

The highway system includes several expressways that radiate out from the *centro* to connect it with the Avenida General Paz, which circles most of the city, thus forming a spokelike pattern. Other main avenues connect the Plaza de Mayo, in the heart of the city, with outlying neighbourhoods. The city is the terminus of every major railway in the country. There are also electric suburban lines connecting the city with the towns of Tigre and Moreno.

The international airport of Ezeiza, Don Torcuato airport, and El Palomar, the military airport, are respectively located outside the city limits in the *partidos* of Esteban Echeverria, Tigre, and Morón. They are connected to the city by expressways. Jorge Newbery, the Buenos Aires City Airport, lies within the Federal District and serves domestic airlines, as well as those that operate to and from neighbouring countries.

ADMINISTRATION AND SOCIAL CONDITIONS

Federal power over the city

Government. The municipal government is headed by an appointed intendant, or mayor, and the elected City Council. The legislative and executive powers of the council and mayor are limited, however, by the centralizing tendency of the federal government, which, of course, has its seat in the Federal District.

By law, the president of the republic actually controls the municipality, and the National Congress is ultimately responsible for legislation pertaining to it. The president appoints the mayor. Over the years the elected city council has functioned only intermittently, and in times of military rule the actual governing of the city often has been assumed directly by the junta in power.

The city territory is divided into administrative districts, some of which coincide approximately with the traditional barrios. Each district has a neighbourhood junta, or council, which deals with neighbourhood problems.

Services. Almost all public-service facilities are controlled by national governmental organizations, functioning independently of the city administration. The municipal government is in charge only of the construction and maintenance of pavements and sidewalks and of parks. Because of population growth and uncontrolled building, public services have been severely strained.

Sanitary facilities—water supply, drainage, sewers—are the responsibility of the General Administration of Sanitary Works of the Nation, which services a major part of Greater Buenos Aires. The large water purification plant in Palermo Park treats water from the Río de la Plata and sends it underground through tunnels to reservoirs in different parts of the city. Electricity is provided by the Company of Electrical Services of Greater Buenos Aires, a government stock company. The National Telecommunications Enterprise operates many centres and provides telephone service for the capital and its surrounding areas. Natural gas is provided for all of Greater Buenos Aires by the State Gas Corporation from oil fields to the north, south, and west.

Health. The city has municipal and private hospitals, as well as numerous dispensaries, pharmacies, and social hygiene centres in the various barrios. A relatively high number of hospital beds are maintained in order to accommodate the many patients from other parts of Argentina who travel to Buenos Aires for specialized treatment.

The waste disposal system includes garbage-burning centres, but most wastes are used to fill lowlands along the Río de la Plata. Reclaimed areas are converted into parks, sports areas, and construction sites.

Education. Argentines are among the most literate peo-

The University of Buenos Aires

ple in the world, and the *porteño* has access to the country's finest educational facilities. The primary and secondary schools are based on the neighbourhood system, and the government runs special national high schools as college preparatory schools. Students in those high schools and others, including numerous schools administered by the Roman Catholic Church or by other religious orders, aim for entrance into the University of Buenos Aires (1821). Despite the repressive years of military dictatorship, this institution has continued to produce outstanding students and teachers. Faculty members have earned Nobel prizes in the medical and scientific fields. The political difficulties suffered by the University of Buenos Aires has encouraged the foundation of several private universities, the most prominent being the University of Belgrano. In addition, the city is the site of the University of Morón, and there are two major universities run by the Roman Catholic Church, the Catholic University of Argentina and Salvador University.

CULTURAL LIFE

Military curtailment of free expression

Buenos Aires' monopoly over the nation's activities is, perhaps, most evident in the capital's cultural life. The plastic arts are always active, with dozens of galleries along and near Calle Florida, just off Avenida Santa Fe, or in elegant, remodeled colonial houses in the San Telmo district. Every modernist movement of note has had its adherents in Buenos Aires, and it is easy to see the world movements in art, letters, theatre, and music reflected in the city. The magnificent opera house, for instance, the Colón Theatre, is one of the major stops for opera stars from all parts of the world. Cultural life in the capital has been peculiarly vulnerable to political change. Military governments, wary of artistic expression, tended to be more restrictive than democratic governments. Newspapers particularly have struggled for free expression during periods of military dominance. Buenos Aires, one of the major publishing centres of Latin America, has a wide variety of newspapers, including several in foreign languages. At least two dailies, *La Prensa* and *La Nación,* are internationally known.

The performing arts. The most important of the city's theatres is the Colón Theatre, which is the opera house as well as the headquarters of the national ballet and the national symphony. The San Martín Municipal Theatre houses three simultaneously available stages as well as an art gallery; the Presidente Alvear Theatre offers performances almost daily. Another theatre of note is the Cervantes National Theatre. Amateur theatre groups are active, and there is an open-air stage in La Boca. Performances of folk songs and dances, known as *peñas,* are extremely popular.

Museums and libraries. The city's museums—several of which are run by the municipal government—house varied collections. The Bernardino Rivadavia Argentine Museum of Natural Sciences has an exceptionally rich fossil collection and operates a scientific institute. The National Museum of Fine Arts contains collections of world masters and of Argentine painters and sculptors, while the National Museum of Decorative Arts houses tapestries and antiques. The Isaac Fernández Blanco Municipal Museum of Hispanic-American Art contains antique silver objects in a replica of a colonial home. Other art collections include Spanish, Italian, and modern works, and there are several historical museums and documentary centres.

The public municipal libraries, distributed throughout the city, are complemented by larger university and institutionally sponsored libraries. The National Library, the city's largest, is now housed in a modern building in Barrio Norte.

Recreation. The city is ringed with *espacios verdes* ("green areas"), which include plazas, parks, and tree-lined boulevards. Two extensive parks that were built on reclaimed floodlands are the Almirante Brown Park, in the Riachuelo Valley, and the Tres de Febrero Park, on the Río de la Plata. Palermo is the oldest park.

Favourite sports

Equestrian sports are enormously popular, especially polo, which has become the national sport. *Pato,* another popular game played on horseback, traces its origin to nomadic pampean Indians. The most popular team sport is football (soccer), which can be seen being played in parks and on fields throughout the urban area. Sports facilities include the Municipal Auto Race Track, the zoological and botanical gardens, the university campuses, the public baths, and various nautical and sports clubs.

(J.M.F.P./Jo.B./J.S.T.)

History

THE EARLY PERIOD

Founding

The city of Buenos Aires was founded twice. It was first founded in 1536 by an expedition led by the Spaniard Pedro de Mendoza, who had been made the first governor general of the Río de la Plata region. That settlement soon fell victim to local Indians and to deficient supplies, and the survivors had to retreat up the river to the fortified settlement of Asunción. Nearly 50 years later Juan de Garay led a more substantial expedition back to the site, and there, at the mouth of the Río Riachuelo, he refounded Nuestra Señora Santa María del Buen Aire in 1580. Huge tracts of land in the environs of the city were granted to members of the expedition, and they began immediately to harvest the pastoral animals that had multiplied since having been left by the original party.

For nearly two centuries Buenos Aires grew at a modest pace. It was a reasonably good port, but it suffered from the rigid organization of the Spanish Empire in America, by which only selected ports could be used for trade. The entire Río de la Plata region was made part of the Viceroyalty of Peru and governed from Lima. Within the Viceroyalty, only Callao, the port near Lima, was granted permission to trade with Spanish merchants. This effectively reduced Buenos Aires to a backwater. Goods from Callao took nearly six months to reach Buenos Aires by oxcart. Any goods the settlers wanted to sell to Spain took that long to reach Callao and another four or six months before they might be shipped from the port to Cádiz. A complete exchange took at least 24 months.

The vast distance separating Buenos Aires from other centres of population in the viceroyalty left the city with only sporadic contact with the administrative authority of the crown. Gradually, the city evolved its own way of life, based on extensive ranching and contraband trade, while the rest of the viceroyalty was focused to one degree or another on the mining enterprises of the Andean region, called Upper Peru (modern Bolivia). A string of settlements was established along the foothills of the Andes to serve the mining region. Their links with the port on the Río de la Plata were of little consequence.

Contraband trade

Instead of suffering from neglect, the *porteños* thrived. In the last quarter of the 17th century and the first half of the 18th, settlements spread rapidly to the northwest along the banks of the Paraná River, a fertile area well irrigated by many streams and small rivers; these were easily navigated by small boats operated by smugglers who reached the many farms and ranches that lined the river. By the beginning of the 18th century, Argentines were exporting thousands of tons of cereals, tens of thousands of cattle hides, and tons of dried beef destined for the plantations of northern Brazil and the Caribbean islands. The British were the principal source of capital and of transportation for this contraband trade.

By the middle of the 18th century, Buenos Aires was a thriving, if still modest commercial entrepôt of nearly 20,000 inhabitants. The houses were built along the narrow earthen streets stretching north from the Riachuelo. The original harbour had become silted up, and the larger boats that now called at the port had to anchor offshore. But the economic success of the region was undeniable, and in 1776, as part of the Bourbon monarchy's broad reform effort, Buenos Aires was named the capital of the new Viceroyalty of the Río de la Plata. The Bourbon monarchs hoped that by expanding their administrative setup in the Americas they would increase tax revenue from the colonies and, at the same time, increase control over the colonies to protect them from the covetous attentions of Spain's rivals, especially the British. Trade out of Buenos Aires was by this time legal, although the crown still attempted to control its flow and pattern. Because the major mining towns of Upper Peru were now within the confines of the viceroyalty, silver was the most valuable export. The city flourished, and, over the last quarter of the 18th and in the early 19th century, the population of the city nearly doubled, from 24,000 in 1778 to 42,500 in 1810. Official trade reflected Buenos Aires' position as the administrative centre of the viceroyalty. Spain became the region's principal trading partner.

The divided elite

Perhaps the most significant result of the administrative reforms of 1776 was that they split the elite into two groups whose economic interests were divergent. One continued to concentrate its energies on the pastoral activities of the city's hinterland and the related trade with Cuba, Brazil, and Great Britain; its interests were more international. The other group was tied economically and administratively to the official activities of the viceroyalty. It was linked to the official bodies, such as the *consulado* (the trade board), that were recognized by the crown and through which the crown attempted to channel all economic activity. This group's interests were more regional.

THE INDEPENDENT CAPITAL

The independent spirit of the city was given a tremendous boost in 1806 and 1807 when local militia forces fought off two attempted invasions by British troops. Neither invasion was a major effort, but the fact that local forces had defeated a British army marked the initial episode in the history of Argentine nationalism. In 1808 when Napoleon invaded Spain and placed his brother Joseph Bonaparte on the throne in Madrid, many *porteños,* like people throughout the empire, reconsidered their ties to the crown. In May 1810 the town council severed ties with Spain and the viceregal government, and on May 25 it declared allegiance to a new ruling junta.

The events of the next decade emphasized the split between the city and the rest of the region. Few residents of the interior were disposed to follow the lead of Buenos Aires, and it was not until 1816, at a congress in Tucumán, that the other provinces declared their independence. A provisional government was created, and Buenos Aires was named capital of the United Provinces of the Río de la Plata. The more distant provinces of the former viceroyalty, Bolivia, Uruguay, and Paraguay, refused to become part of a new nation dominated by the port city, however. For nearly 30 years, the provinces were held together in an uneasy truce, called federalism, in which the city of Buenos Aires and its hinterland exercised whatever central authority existed in the new nation; the interior provinces were allowed to go their own way.

The United Provinces of the Río de la Plata

Ironically, it was the interior provinces that suffered most from this arrangement, and in 1851 they mounted a coalition that attempted to change the balance of power by force. They succeeded in ousting the *porteño* leader, Juan Manuel de Rosas, but were unable to reorganize the nation in an effective manner. A decade later, the *porteños,* led by the military leader Bartolomé Mitre, imposed their will on the interior through a strong government centred in Buenos Aires. This development was recognized officially when the city was made the federal capital in 1880.

The political struggles between the *porteños* and the interior became more intense after 1850 because the stakes became greater. Dramatic changes in the European market as a result of industrialization and the transformation of capitalism, together with significant advances in technology, made exploitation of the fertile plains of Argentina economically viable. All that was required was the labour to work the land and the capital to pay for the transportation of products to the ports. For the most part, the capital came from Britain. The bulk of the labour came from Spain and Italy. In little more than a generation the land was transformed. By the beginning of World War I, Argentina had become one of the world's principal exporters of agricultural products.

The economic change in the countryside led to three fundamental changes in the character of the city. First, the population changed. Immigrants who had been attracted to the country with the hope of settling on the land found it impossible to buy land and they migrated back to the city. At the same time, the need for new port facilities and service activities related to increasing exports created a demand for labour. Those newcomers, mainly from Spain and Italy, but also from eastern Europe and Germany, jammed into the older houses on the south side of the city, pushing the middle-class residents north across the Avenida de Mayo. Because many jobs were in the port and in the slaughterhouses on the outskirts of the city, the newcomers also pushed south across the Riachuelo into adjoining counties.

The European immigrants

The second significant change in the city was the massive amount of wealth that came into the hands of individuals and to the state. The former built great mansions, modeled after French châteaus. These mansions today house government ministries or the embassies of foreign governments. At the time, they were the international symbol of vast wealth. In Paris, at the turn of the century, a common phrase was "to be as rich as an Argentine." These mansions were constructed in the Barrio Norte, many around the Plaza San Martín, at the northern end of the Calle Florida, or close to the Avenida Santa Fe.

The Paris of South America

The third major transformation was in the physical layout of the city. The owners of the mansions, members of the ruling elite, decided that they would transform Buenos Aires into the Paris of South America. As part of the preparations for the May 1910 celebration of the centennial of the first declaration of independence, the City Council decided to build a subway system and a network of broad avenues radiating out from the city centre, in frank imitation of the urban reforms imposed on Paris by Napoleon III. The plan called for the construction of broad avenues every four blocks, running east and west, and every 10 blocks, running north and south, and for construction of what would be claimed as the broadest avenue in the world, patterned on the Champs-Élysées. Named 9 de Julio after the official national day of independence, that block-wide swath was cut through the city in the 1930s and opened officially in October 1937.

The infrastructure put into place in the years before World War I endures to the present. The broad avenues in the core of the city, most of them carved out during the 1920s and '30s, continue to carry the burden of vehicular traffic, and only one of the five subway lines was built after World War II. Although the infrastructure of the city is severely strained, the urban structure still adheres to the form conceived by the generation of leaders born in the 19th century.

THE MODERN CITY

Since the mid-20th century, three important developments have affected the character of the city. The first was the virtual halting of international immigration after 1930. The demand for labour in the metropolitan area was met thereafter, in large measure, by migrants from the interior—Argentines of mestizo ancestry, whose presence created conflicts with the *porteños.* These people came from northwestern Argentina or from the neighbouring countries of Uruguay, Paraguay, and Bolivia.

Migration of the mestizos

The second development stemmed from the movement of the mestizos, who, finding the inner city slums too crowded, settled on unoccupied land in the suburbs near the manufacturing establishments that provided many jobs. The mestizos founded the first shantytowns, the *villas miserias.* The racial difference of these newer migrants to the city added a nuance of bitterness to the social conflicts that characterized the development of industrial capitalism in the metropolitan area in the 1930s, '40s, and '50s. Many of the newcomers would become staunch followers of the populist leader Juan D. Perón, who came to power in a military coup in 1943 and served as president from 1946 to 1955 and again from 1973 until his death in 1974.

The third important development in the second half of the century was the replacement of the tramway and train by the automobile and the colectivo as the dominant modes of transportation in the city. Unlike many other large Western cities, Buenos Aires is not yet ringed by a network of superhighways. A complete network was planned after World War II, but economic and political difficulties have prevented their construction. By the late

20th century the existing network of streets was saturated with vehicular traffic, and the need to improve other modes of transportation seemed imperative as traffic jams and gridlock added to the more frustrating characteristics of contemporary Buenos Aires. Some progress was made in 1979 when the Metropolitan Railroad was established to cope with the capital's traffic problems. City and suburban authorities were coordinated, and in succeeding years plans to expand rail lines and, employing Japanese companies, to electrify them, were implemented. New Japanese-built railroad cars began replacing older cars on the line.

Urban problems in the late 20th century

The new underclass in the city, the recent migrants, have complicated the city's problems. These peoples are undereducated and lack employment skills, and it has been difficult for the service-oriented economy to absorb and employ them. The government, no matter how willing, has had little success in finding the resources to help the newcomers raise their standard of living. Dealing with massive urban poverty and its accompanying social and financial problems was a major concern of the national and local governments in the final decades of the 20th century.

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(J.S.T.)

Building Construction

Building construction is an ancient human activity. It began with the purely functional need for a controlled environment to moderate the effects of climate. Constructed shelters were one means by which human beings were able to adapt themselves to a wide variety of climates and become a global species.

Human shelters were at first very simple and perhaps lasted only a few days or months. Over time, however, even temporary structures evolved into such highly refined forms as the igloo. Gradually more durable structures began to appear, particularly after the advent of agriculture, when people began to stay in one place for long periods. The first shelters were dwellings, but later other functions, such as food storage and ceremony, were housed in separate buildings. Some structures began to have symbolic as well as functional value, marking the beginning of the distinction between architecture and building.

The history of building is marked by a number of trends. One is the increasing durability of the materials used. Early building materials were perishable, such as leaves, branches, and animal hides. Later, more durable natural materials—such as clay, stone, and timber—and, finally, synthetic materials—such as brick, concrete, metals, and plastics—were used. Another is a quest for buildings of ever greater height and span; this was made possible by the development of stronger materials and by knowledge of how materials behave and how to exploit them to greater advantage. A third major trend involves the degree of control exercised over the interior environment of buildings: increasingly precise regulation of air temperature, light and sound levels, humidity, odours, air speed, and other factors that affect human comfort has been possible. Yet another trend is the change in energy available to the construction process, starting with human muscle power and developing toward the powerful machinery used today.

The present state of building construction is complex. There is a wide range of building products and systems which are aimed primarily at groups of building types or markets. The design process for buildings is highly organized and draws upon research establishments that study material properties and performance, code officials who adopt and enforce safety standards, and design professionals who determine user needs and design a building to meet those needs. The construction process is also highly organized; it includes the manufacturers of building products and systems, the craftsmen who assemble them on the building site, the contractors who employ and coordinate the work of the craftsmen, and consultants who specialize in such aspects as construction management, quality control, and insurance.

Building construction today is a significant part of industrial culture, a manifestation of its diversity and complexity and a measure of its mastery of natural forces, which can produce a widely varied built environment to serve the diverse needs of society. This article first traces the history of building construction, then surveys its development at the present time. For treatment of the aesthetic considerations of building design, see ARCHITECTURE, THE ART OF. For further treatment of historical development, see AFRICAN ARTS; ARCHITECTURE, THE HISTORY OF WESTERN; CENTRAL ASIAN ARTS; EAST ASIAN ARTS; EGYPTIAN ARTS AND ARCHITECTURE; ISLĀMIC ARTS; SOUTH ASIAN ARTS; and SOUTHEAST ASIAN ARTS.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 733, and the *Index.*

The article is divided into the following sections:

The history of building construction

PRIMITIVE BUILDING: THE STONE AGE

The hunter-gatherers of the late Stone Age, who moved about a wide area in search of food, built the earliest temporary shelters that appear in the archaeological record. Excavations at a number of sites in Europe dated to before 12,000 BC show circular rings of stones that are believed to have formed part of such shelters. They may have braced crude huts made of wooden poles or have weighted down the walls of tents made of animal skins, presumably supported by central poles.

Elements of shelter

A tent illustrates the basic elements of environmental control that are the concern of building construction. The tent creates a membrane to shed rain and snow; cold water on the human skin absorbs body heat. The membrane reduces wind speed as well; air over the human skin also promotes heat loss. It controls heat transfer by keeping out the hot rays of the sun and confining heated air in cold weather. It also blocks out light and provides visual privacy. The membrane must be supported against the forces of gravity and wind; a structure is necessary. Membranes of hides are strong in tension (stresses imposed by stretching forces), but poles must be added to take compression (stresses imposed by compacting forces). Indeed, much of the history of building construction is the search for more sophisticated solutions to the same basic problems that the tent was set out to solve. The tent has continued in use to the present. The Saudi Arabian goats' hair tent, the Mongolian yurt with its collapsible wooden frame and felt coverings, and the American Indian tepee with its multiple pole supports and double membrane are more refined and elegant descendants of the crude shelters of the early hunter-gatherers.

The agricultural revolution, dated to about 10,000 BC, gave a major impetus to building construction. People no longer traveled in search of game or followed their herds but stayed in one place to tend their fields. Dwellings began to be more permanent. Archaeological records are scanty, but in the Middle East are found the remains of whole villages of round dwellings called tholoi, whose walls are made of packed clay; all traces of roofs have disappeared. In Europe, tholoi were built of dry-laid stone with domed roofs; there are still surviving examples (of more recent construction) of these beehive structures in the Alps (Figure 1). In later Middle Eastern tholoi a rectangular antechamber or entrance hall appeared, attached to the main circular chamber—the first examples of the rectangular plan form in building. Still later the circular form was dropped in favour of the rectangle as dwellings were divided into more rooms and more dwellings were placed together in settlements. The tholoi marked an important step in the search for durability; they were the beginning of masonry construction.

Evidence of composite building construction of clay and wood, the so-called wattle-and-daub method, is also found in Europe and the Middle East. The walls were made of small saplings or reeds, which were easy to cut with stone tools. They were driven into the ground, tied together laterally with vegetable fibres, and then plastered over with wet clay to give added rigidity and weatherproofing. The roofs have not survived, but the structures were probably covered with crude thatch or bundled reeds. Both round and rectangular forms are found, usually with central hearths.

Early frame construction

Heavier timber buildings also appeared in Neolithic cultures, although the difficulties of cutting large trees with stone tools limited the use of sizable timbers to frames. These frames were usually rectangular in plan, with a central row of columns to support a ridgepole and matching rows of columns along the long walls; rafters were run from the ridgepole to the wall beams. The lateral stability of the frame was achieved by burying the columns deep in the ground; the ridgepole and rafters were then tied to the columns with vegetable fibres. The usual roofing material was thatch: dried grasses or reeds tied together in small bundles, which in turn were tied in an overlapping pattern to the light wooden poles that spanned between the rafters. Horizontal thatched roofs leak rain badly, but, if they are placed at the proper angle, the rainwater runs off before it has time to soak through. Primitive builders soon determined the roof pitch that would shed the water but not the thatch. Many types of infill were used in the walls of these frame houses, including clay, wattle and daub, tree bark (favoured by American Woodland Indians), and thatch. In Polynesia and Indonesia, where such houses are still built, they are raised above the ground on stilts for security and dryness; the roofing is often made of leaves and the walls are largely open to allow air movement for natural cooling. Another variation of the frame was found in Egypt and the Middle East, where timbers were substituted for bundles of reeds.

Werner Blaser

Figure 1: Corbeled stone tholos in the Swiss Alps, 20th century.

BRONZE AGE AND EARLY URBAN CULTURES

It was the cultures of the great river valleys—including the Nile, the Tigris and Euphrates, the Indus, and the Huang Ho—with their intensive agriculture based on irrigation, that developed the first communities large enough to be called cities. These cities were built with a new building technology, based on the clay available on the riverbanks. The packed clay walls of earlier times were replaced by those constructed of prefabricated units: mud bricks. This represented a major conceptual change from the free forms of packed clay to the geometric modulation imposed by the rectangular brick, and the building plans too became strictly rectangular.

Bricks were made from mud and straw formed in a four-sided wooden frame, which was removed after evaporation had sufficiently hardened the contents. The bricks were then thoroughly dried in the sun. The straw acted as reinforcing to hold the brick together when the inevitable shrinkage cracks appeared during the drying process. The bricks were laid in walls with wet mud mortar or sometimes bitumen to join them together; openings were apparently supported by wooden lintels. In the warm, dry climates of the river valleys, weathering action was not a major problem, and the mud bricks were left exposed or covered with a layer of mud plaster. The roofs of these early urban buildings have disappeared, but it seems likely that they were supported by timber beams and were mostly flat, since there is little rainfall in these areas. Such mud brick or adobe construction is still widely used in the Middle East, Africa, Asia, and Latin America.

Appearance of fired brick

Later, about 3000 BC in Mesopotamia, the first fired bricks appeared. Ceramic pottery had been developing in these cultures for some time, and the techniques of kiln-firing were applied to bricks, which were made of the same clay. Because of their cost in labour and fuel, fired bricks were used at first only in areas of greater wear, such as pavements or the tops of walls subject to weathering. They were used not only in buildings but also to build sewers to drain wastewater from cities. It is in the roofs of these underground drains that the first surviving true arches in brick are found, a humble beginning for what would become a major structural form. Corbel vaults and domes made of limestone rubble appeared at about the same time in Mesopotamian tombs (Figure 2). Corbel vaults are constructed of rows of masonry placed so that each row projects slightly beyond the one below, the two opposite walls thus meeting at the top. The arch and the vault may have been used in the roofs and floors of other buildings, but no examples have survived from this period. The well-developed masonry technology of Mesopotamia was used to build large structures of great masses of brick, such as the temple at Tepe Gawra and the ziggurats at Ur and Borsippa (Birs Nimrud), which were up to 26 metres (87 feet) high. These symbolic buildings marked the beginnings of architecture in this culture.

The development of bronze, and later iron, technology

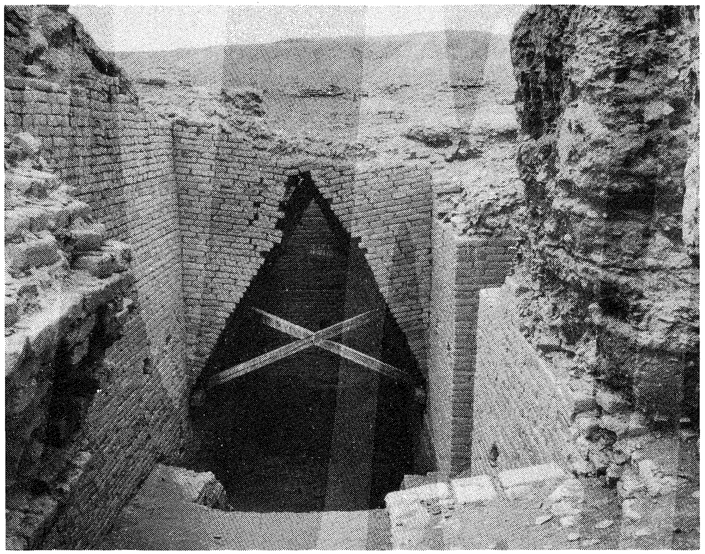

Figure 2: Brick walls and corbel vault at the entrance to the tomb chamber of Ur-nammu in the royal mausoleum at Ur, late 3rd millennium BC.

Hirmer Verlag, Munich

in this period led to the making of metal tools for working wood, such as axes and saws. Less effort was thus required to fell and work large trees. This led in turn to new developments in building technics; timbers were cut and shaped extensively, hewed into square posts, sawed into planks, and split into shingles. Log cabin construction appeared in the forested areas of Europe, and timber framing became more sophisticated. Although the excavated remains are fragmentary, undoubtedly major advances were made in timber technology in this period; some of the products, such as the sawed plank and the shingle, are still used today.

STONE CONSTRUCTION IN EGYPT

Like the other great river valley cultures, Egypt built its cities with mud brick; fired brick did not appear there until Roman times. Timber was used sparingly, for it was never abundant. It was used mainly in roofs, where it was heavily supplemented by reeds. Only a few royal buildings were built with full timber frames.

It was against this drab background of endless mud brick houses that a new technology of cut-stone construction emerged in the temples and pyramids of the 4th dynasty (*c.* 2575–*c.* 2465 BC). Egypt, unlike Mesopotamia or the Indus valley, had excellent deposits of stone exposed above ground; limestone, sandstone, and granite were all available. But the extracting, moving, and working of stone was a costly process, and the quarrying of stone was a state monopoly. Stone emerged as an elite construction material used only for important state buildings.

The Egyptians developed cut stone for use in royal mortuary buildings not only for its strength but also for its durability. It seemed the best material to offer eternal protection to the pharaoh's *ka,* the vital force he derived from the sun-god and through which he ruled. Thus stone had both a functional and symbolic significance.

Djoser's complex at Ṣaqqārah

Within the long tradition of brick masonry, stone construction appeared abruptly, with little transition. The brick mastaba tombs of the early kings and nobles suddenly gave way to the stone technics of King Djoser's ceremonial complex at Ṣaqqārah, the construction of which is associated with his adviser and builder Imhotep. It is a structure of somewhat curious and uncertain forms but of great elegance in execution and detail. It consists mostly of massive limestone walls that enclose a series of interior courtyards. The walls have convoluted surfaces, which recall the mastaba tombs, with dummy doors, and there are even whole dummy buildings of solid stone. The complex has a large entrance hall with a roof supported by massive stone lintels that rest on rows of short wing walls projecting from the enclosing walls. There are no free-standing columns, but incipient fluted columns appear at the ends of the wing walls and engaged $^3/_4$-columns project from the walls of the courtyards. The complex also contains the first pyramid, created from successively smaller mastabas. All these elements are built of small stones, which could be handled by one or two men. It represents a technology that was already highly developed, involving elaborate methods of quarrying, transporting, and working stone.

Quarrying methods

The construction process began at the quarries. Most of them were open-faced, although in some cases tunnels were extended several hundred metres into cliffs to reach the best quality stone. For extracting sedimentary rock, the chief tool was the mason's pick with a 2.5-kilogram metal head and a 45-centimetre haft. With these picks vertical channels as wide as a man were cut around rectangular blocks, exposing five faces. The final separation of the sixth face was accomplished by drilling rows of holes into the rock with metal bow drills. Wooden wedges were driven into the holes to fill them completely. The wedges were doused with water, which they absorbed and which caused them to expand, breaking the stone free from its bed. In the extraction of igneous rock such as granite, which is much harder and stronger than limestone, the mason's pick was supplemented by balls of dolerite weighing up to 5 kilograms, which were used to break the rock by beating and pounding. Granite was also drilled and sawed with the help of abrasives, and expanding wooden wedges were used in splitting.

The Egyptians were able to move blocks weighing up to 1,000,000 kilograms from quarries to distant building sites. This was an amazing accomplishment, as their only machinery was levers and crude wooden sledges worked by masses of men and draft animals. There were no wheeled vehicles before 1500 BC, and they were never widely used in building. Most quarries were near the Nile, however, and boats were also extensively used in transporting stone.

At the building site the rough stones were precisely finished to their final forms, with particular attention to their exposed faces. This was done with metal chisels and mallets; squares, plumb bobs, and straightedges were used to check the accuracy of the work. These tools remained standard until the 19th century. After the first appearance of small stones at Ṣaqqārah, their size began to increase until they attained the cyclopean scale usually associated with Egyptian masonry at about the time of the building of the pyramids. In spite of the heavy loads that stone structures created, foundations were of a surprisingly shoddy and improvised character, made of small blocks of poor quality stone. Not until the 25th dynasty (*c.* 750–656 BC) were important buildings placed on a below-grade (underground) platform of masonry several metres thick.

The Egyptians possessed no lifting machinery to raise stones vertically. It is generally thought that the laying of successive courses of masonry was accomplished with earth or mud brick ramps, over which the stones were dragged to their places in the walls by animal and human muscle power. Later, as the ramps were removed, they served as platforms for the masons to apply the final finishes to the stone surfaces. The remains of such ramps can still be seen at unfinished temples that were begun in the Ptolemaic period. The stones were usually laid with a bed of mortar made of gypsum, sand, and water, which perhaps acted more as a lubricant to push the stone into place than as a bonding agent. There was also limited use of metal dovetail anchors between blocks.

Stone frame construction

The great Pyramids of Giza, the tallest of which rose to a height of 147 metres (481 feet), are a marvelous technological achievement, and their visual impact is stunning even today; it was not until the 19th century that taller structures would be built. But they also represent a dead end in massive stone construction, which soon moved in the direction of lighter and more flexible stone frames and the creation of larger interior spaces. The free-standing stone column supporting stone beams appeared for the first time in the royal temples associated with the pyramids of about 2600 BC (Figure 3). Square granite columns carrying heavy granite lintels spanned three to four metres (10 to 13 feet); the spaces between the lintels were roofed by massive granite slabs. In these structures the abstract

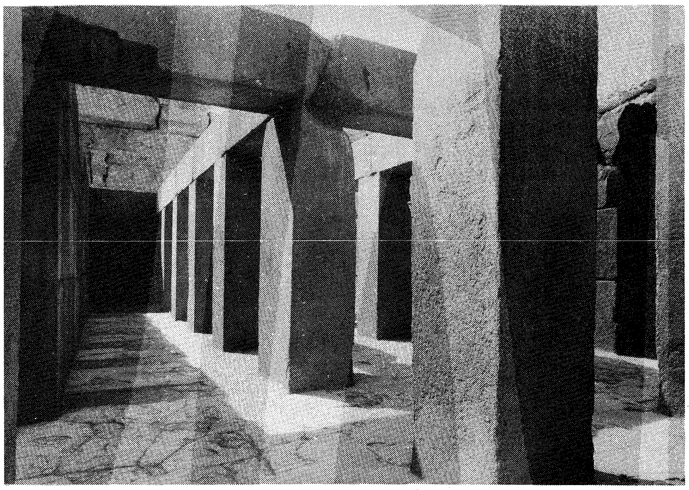

Figure 3: Stone frame showing post-and-lintel construction, Khafre's Valley Temple, Giza, *c.* 2600 BC.
Hirmer Verlag, Munich

notion of the timber frames of the early royal buildings was translated into stone.

Although stone is more durable than timber, it is quite different in structural strength. Stone is much stronger in compression than timber but is weaker in tension. For this reason, stone works well for columns, which could be made very high—for example, 24 metres (80 feet) in the great temple of Amon-Re at Karnak. But stone lintels spanning between columns are limited by the tension they develop on their bottom surfaces; their maximum span is perhaps five metres (16 feet). Thus, for longer spans, another structural form was needed to exploit the higher compressive strength of stone. But the arch, which could span a longer distance in compression, remained confined to the sewers and to the underground roofs of the tombs of minor officials. So, perhaps with the image of the timber building frame still strong in their minds, the Egyptian masons were content to explore the limitations of the analogous stone frame in a series of great temples built during the New Kingdom (1539–1075 BC) at Karnak and Luxor, culminating in the elegant loggias of Queen Hatshepsut's temple at Dayr al-Baḥrī. The paradigm of the stone-frame temple that they established would endure to the end of the classical world.

GREEK AND HELLENISTIC CULTURES

Use of the Egyptian stone frame diffused throughout the eastern Mediterranean after 1800 BC, and the cultures of mainland Greece were particularly attracted to it. In the Greek world of the Aegean and southern Italy, many stone-frame temples were built; some have survived to the present day in various states of preservation. They were built largely of local marble or limestone; there was no granite for huge monoliths. The basic technology was little changed from that of Egypt; the major difference was in the labour force. There were no state-mobilized masses of unskilled workers to move huge stones; there were instead small groups of skilled masons who worked independently. The building accounts of the Parthenon show that each column was built under a separate contract with a master mason. There was certainly lifting machinery for handling the blocks, although its precise description is unknown; the concealed faces of stones still have grooves and holes that engaged the ropes used to lift them into place. Metal cramps and dowels were introduced for joining stones together; mortar was almost never used. There was some experimentation with iron beams to reinforce longer spans in stone, but the maximum remained about five to six metres (16 to 20 feet). Longer spans were achieved with timber beams supported by the stone frame; the solid stone roof slabs of the great Egyptian temples could not be duplicated.

Much of the mason's effort was concentrated on the refinements of detail and optical corrections for which Greek architecture is justly famous. This same sense is also seen in the first surviving construction drawings, which were made on the unfinished surfaces of the stone walls of the Temple of Didyma. Such drawings would normally have been erased during the final finishing of the wall surfaces, and those at Didyma survived because the temple was never completed. The drawings show how the masons developed the final profiles of columns and moldings—a rare glimpse of the design processes of builders before the days of pencil and paper.

First use of construction drawings

In contrast to stone technology, which remained largely unchanged from Egyptian methods, clay masonry underwent considerable development. Although mud brick remained standard for dwellings, fired brick was more widely used and began to be laid with lime mortar, a technique borrowed from stone construction. Glazed brick also appeared in this period, particularly outside the Greek world among the Babylonians and Persians, who made considerable use of it in royal palaces. A fine surviving example is the Ishtar Gate of the Palace of Nebuchadrezzar at Babylon, with a true arch spanning 7.5 metres (25 feet) and dated to 575 BC. Another major innovation was the fired clay roof tile. This was much more waterproof than thatch, and tile roofs could have the lower pitch characteristic of Greek temples. Hollow terra-cotta blocks for wall ornaments also appeared about this time, probably derived from the highly advanced pottery industry, which routinely made fired clay vessels more than one metre long.

Although stone technology remained confined to the trabeated (column-and-beam, or post-and-lintel) frame, there were a few structures that hinted at future developments. Perhaps the most spectacular building achievement of the age was the Pharos of Alexandria, the great lighthouse built for Ptolemy II in the 3rd century BC. It was a huge stone tower nearly as high as the Great Pyramid but much smaller at the base—perhaps 30 metres (100 feet) square. Within this mass of masonry was a complex system of ramps over which pack animals carried fuel for the beacon at the top. The Pharos was the first high-rise building, but the limitations of masonry structures and the lack of a rapid way of moving people vertically precluded any further development of tall buildings until the 19th century. The Pharos remained the only example of this type long after it was demolished by the Arabs beginning in the 7th century AD.

Another example of a new stone technology that was tried but not pursued further by the Greeks was the underground tombs of Mycenae, built about 1300 BC. These tombs have main chambers enclosed by pointed domes of corbeled stone construction, about 14 metres (47 feet) in diameter and 13 metres (43 feet) high. Crude versions of the corbel dome had appeared earlier in Mesopotamian tombs and the tholoi of Neolithic Europe, but in Mycenae the technics were refined and enlarged in scale. A corbel dome or arch does not develop the high compressive forces that characterize true arches and domes, which are built of radial segments of stone or brick. Thus it does not take full advantage of the great compressive strength of stone and cannot span long distances; 14 metres is near the upper limit. Greek masons did not choose to explore this type of structure; their buildings remained largely concerned with exterior forms. The Roman builders who followed them, however, exploited masonry to its full potential and created the first great interior spaces.

ROMAN ACHIEVEMENTS

It was from the Etruscans, who lived in the northern part of Italy, that the Romans derived much of their early building technology. The Etruscans, probably influenced by a few rare Greek examples in southern Italy, developed the true arch in stone. A late specimen of the 3rd century BC is the Porta Marzia, an arched city gateway with a span of about six metres (20 feet), in Perugia. The Etruscans also had a highly developed terra-cotta technology and made excellent fired bricks.

Masonry construction. The Romans adopted Etruscan stone construction based on the arch and built many spectacular examples of what they called *opus quadratum,* or structures of cut stone blocks laid in regular courses. Most of these were public works in conquered provinces, such

Opus quadratum

as the late 1st-century-BC Pont du Gard, a many-arched bridge and aqueduct spanning 22 metres (72 feet) near Nîmes, in France, or the fine bridge over the Tagus River at Alcântara in Spain, with a span of almost 30 metres (100 feet), built about AD 110. Oddly enough, such long spans in stone were never applied to buildings. The surviving Roman buildings with stone arches or vaults have typical spans of only four to seven metres (15 to 25 feet); small stone domes with diameters of four to nine metres were built in Roman Syria. Such arches and domes imply the existence of sophisticated timber formwork to support them during construction, as well as advanced lifting machinery, but there are no extant records of either. Many of these structures survived the fall of the empire, and they became models for the revival of stone construction in medieval Europe, when masons again sought to build "in the Roman manner." The Romans also inherited the trabeated stone frame from the Greeks of southern Italy and continued to build temples and other public buildings with this type of construction into the 3rd century AD.

Brickmaking, particularly in the region of Rome itself, became a major industry and finally, under the empire, a state monopoly. Brick construction was cheaper than stone due to the economies of scale in mass production and the lower level of skill needed to put it in place. The brick arch was adopted to span openings in walls, precluding the need for lintels. Mortar was at first the traditional mixture of sand, lime, and water, but, beginning in the 2nd century BC, a new ingredient was introduced. The Romans called it *pulvis puteoli* after the town of Puteoli (modern Pozzuoli), near Naples, where it was first found; the material, formed in Mount Vesuvius and mined on its slopes, is now called pozzolana. When mixed with lime, pozzolana forms a natural cement that is much stronger and more weather-resistant than lime mortar alone and that will harden even under water. Pozzolanic mortars were so strong and cheap, and could be placed by labourers of such low skill, that the Romans began to substitute them for bricks in the interiors of walls; the outer wythes of bricks were used mainly as forms to lay the pozzolana into place. Finally, the mortar of lime, sand, water, and pozzolana was mixed with stones and broken brick to form a true concrete, called *opus caementicium.* This concrete was still used with brick forms in walls, but soon it began to be placed into wooden forms, which were removed after the concrete had hardened.

Concrete as a building material

One of the earliest surviving examples of this concrete construction is the Temple of the Sybil (or Temple of Vesta) at Tivoli, built during the 1st century BC. This temple has a circular plan with a peristyle of stone columns and lintels around the outside, but the wall of the circular cella, or sanctuary room, inside is built of concrete—an uneasy confrontation of new and traditional forms of construction. An early large-scale example in Rome itself of brick-faced concrete is the plain rectangular walls of the Camp of the Praetorian Guard, built by Sejanus in AD 21–23. But the possibilities of plastic form suggested by this initially liquid material, which could easily assume curved shapes in plan and section, soon led to the creation of a series of remarkable interior spaces, spanned by domes or vaults and uncluttered by the columns required by trabeated stone construction, that showed the power of the imperial state. The first of these is the octagonal domed fountain hall of Nero's Golden House (AD 64–68), which is about 15 metres (50 feet) in diameter with a large circular opening, or oculus, in the top of the dome. The domed form was rapidly developed in a series of imperial buildings that culminated in the emperor Hadrian's Pantheon of about AD 118–128 (Figure 4). This huge circular structure was entered from a portico of stone columns and was surmounted by a dome 43.2 metres (142 feet) in diameter, lighted by an oculus at the top. The walls supporting the dome are of brick-faced concrete six metres (20 feet) thick lightened at intervals by internal recesses; the dome is of solid concrete 1.5 metres (five feet) in average thickness and rising 43.2 metres above the floor. This magnificent structure has survived in good condition to modern times; the diameter of its circular dome remained unsurpassed until the 19th century.

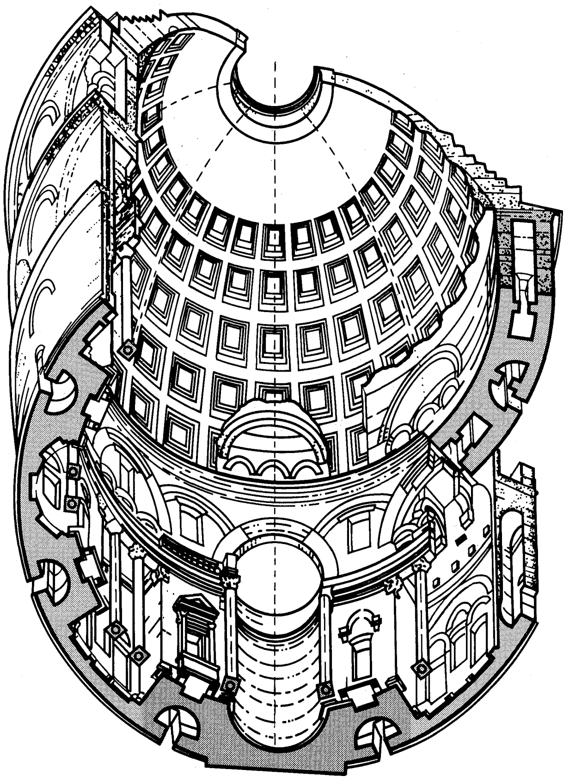

Figure 4: Axonometric sectional view of the Pantheon, Rome.
Adapted from J.B. Ward-Perkins, *Roman Imperial Architecture* (1981), Viking-Penguin, Inc. Reproduced by permission of Penguin Books Ltd.

Two large fragments of great concrete cross-vault buildings still survive from the late empire. The first of these is a portion of the Baths of Diocletian (*c.* 298–306) with a span of 26 metres (85 feet); it was converted into the church of Santa Maria degli Angeli by Michelangelo in the 16th century. The other is the Basilica of Maxentius (AD 307–312), also with a span of 26 metres. All of these buildings contained stone columns, but they were purely ornamental and could have been removed at will. The brick-faced concrete walls were left exposed on the exteriors, but the interiors were lavishly decorated with a veneer of thin slabs of coloured stone held in place by metal fasteners that engaged slots cut in the edges of the slabs, a technique still used in the 20th century. These and other great Roman public spaces spanned by concrete domes and vaults made a major advance in scale over the short spans of the stone frame.

In the late empire, concrete technology gradually disappeared, and even brickmaking ceased in western Europe. But significant developments in brick technology continued in the eastern Roman world, where the achievements of earlier periods in concrete were now duplicated in brickwork. The tomb of the emperor Galerius (now the Church of St. George) of about AD 300 at Salonika in Greece has a brick dome 24 metres (80 feet) in diameter. It probably was the model for the climactic example of late Roman building, the great church of Hagia Sophia (532–537) in Constantinople, which features a central dome spanning 32.6 metres (107 feet). Even Rome's great enemies, the Sāsānian Persians, built a large brick-vaulted hall in the palace at Ctesiphon (usually identified with Khosrow I [mid-6th century] but probably a 4th-century structure) with a span of 25 metres (82 feet) by borrowing Roman methods. These late brick structures were the last triumphs of Roman building technology and would not be equaled for the next 900 years.

Timber and metal construction. The Romans also made major advances in timber technology. Reliefs on Trajan's Column show the timber lattice truss bridges used by Roman armies to cross the Danube. The truss, a hollowed-out beam with the forces concentrated in a triangulated network of linear members, was apparently a Roman in-

Invention of the truss

vention. No evidence of their theoretical understanding of it exists, but nevertheless they were able to master the design of trusses in a practical way. A fine example is the Basilica of Constantine at Trier (AD 297–299), where timber king-post roof trusses (triangular frames with a vertical central strut) span a hall 23 metres (75 feet) wide; the present roof is a restoration, but the original must have been similar.

The notion of the truss was extended from timber to metal. Bronze trusses, running over three spans of about nine metres (30 feet) each, supported the roof of the portico of the Pantheon. The choice of bronze was probably made more for durability than strength, because Pope Urban VIII was able to remove this bronze work in 1625 (to melt it down for cannon) and replace it with timber trusses. The truss remained an isolated achievement of Roman building that would not be equaled until the Renaissance.

Metals were used extensively in Roman buildings. In addition to bronze trusses, the Pantheon had bronze doors and gilded bronze roof tiles. Lead was another material introduced by the Romans for roofing; it was waterproof and could be used with very low pitches.

Building support systems. Perhaps the most important use of lead was for pipes to supply fresh water to buildings and to remove wastewater from them (the word plumbing comes from the Latin *plumbum,* which means lead). The Romans provided generous water supplies for their cities; all of the supply systems worked by gravity and many of them used aqueducts and syphons. Although most people had to carry their water from public fountains, there was limited distribution of water to public buildings (particularly baths) and some private residences and apartment houses; private and semiprivate baths and latrines became fairly common. The wastewater drainage system was limited, with no treatment of sewage, which was simply discharged into a nearby river. But even these fairly modest applications of public sanitation far exceeded those of previous cultures and would not be equaled until the 19th century.

Another material that the Romans applied to buildings was glass, which had been developed by the Egyptians who used it only for jewelry and small ornamental vessels. The Romans devised many kinds of coloured glass for use in mosaics to decorate interior surfaces. They also made the first clear window glass, produced by blowing glass cylinders that were then cut and laid flat. Seneca (*c.* 4 BC–AD 65) described the sensation caused by the appearance of glazed sun porches in the villas near Rome. Although no Roman glass installations have survived, glass apparently became fairly common in public buildings and was even used in middle-class apartment houses in the capital.

Hypocaust heating

In most Roman buildings, the central open fire remained the major source of heat—as well as annoying smoke—although the use of charcoal braziers made some improvement. A major innovation was the development of hypocaust, or indirect radiant, heating, by conducting heated air through flues in floors and walls. The heated masonry radiated a pleasantly uniform warmth, and smoke was eliminated from occupied spaces; the same method was used to heat water for baths. The Basilica of Constantine at Trier has a well-preserved example of hypocaust heating, where the stone slabs of the floor are supported on short brick columns, creating a continuous heating plenum beneath it.

ROMANESQUE AND GOTHIC

The disappearance of Roman power in western Europe during the 5th century led to a decline in building technology. Brickmaking became rare and was not revived until the 14th century. Pozzolanic concrete disappeared entirely, and it would not be until the 19th century that man-made cements would equal it. The use of domes and vaults in stone construction was also lost. Building technics fell to Iron Age levels, exemplified by log construction, packed clay walls, mud brick, and wattle and daub.

Advanced building technologies were developing in China in this same period during the Sui (581–618) and T'ang (618–907) dynasties. In the 3rd century BC the completion of the Great Wall, about 6,400 kilometres (4,000 miles) in length and following a sinuous path along the contours of rugged terrain, had demonstrated remarkable achievements in masonry technology, logistics, and surveying methods. The An-Chi Bridge, built about AD 610 in Hopei province, had a stone arch with a span of 37.5 metres (123 feet), far exceeding the Roman bridge at Alcántara. Extensive work was also done in the development of heavy timber framing (primarily for temples), and stone tower pagodas up to 60 metres (200 feet) high were built; fired brick was also widely used. These elements of Chinese building technology set a high standard of quality that would be maintained until the 19th century.

Stone construction. Beginning in the 9th century there were the first stirrings of the revival of stone construction in Europe; the Palatine Chapel of Charlemagne at Aachen (consecrated 805), with its octagonal segmented dome spanning 14.5 metres (47 feet), is an early example of this trend. But the Romanesque style, building "in the Roman manner" with stone arches, vaults, and domes to span interior spaces, did not really begin until the later part of the 11th century. Vaults reappeared in such structures as the cathedral of Santiago de Compostela in Spain (begun 1078) and Saint Sernin at Toulouse (begun 1080). The cross vault raised on columns was seen again at Speyer Cathedral (1030–65, reconstructed *c.* 1082–1137) and Durham Cathedral (1093–1133), and the domes of St. Mark's Basilica in Venice (late 11th century) and the cathedral of Saint-Front in Périgueux (1120–1150) marked the recovery of the complete range of Roman structural forms. All these buildings were built by the Roman Catholic church, which had spread its influence throughout western Europe in this period. One contemporary chronicler wrote that the earth seemed to be "clothing itself with a white robe of churches," white because they were new and built of stone. From 1050 to 1350 more stone was quarried in France alone than in the whole history of ancient Egypt—enough to build 80 cathedrals, 500 large churches, and tens of thousands of parish churches. The great building campaign of medieval times has been called the "cathedral crusade," an equally impassioned counterpart of the great military adventures to recover the Holy Land. This vast undertaking required many masons, who worked as free craftsmen, organizing themselves into societies or guilds. They oversaw the quarrying of stone, supervised the process of apprenticeship by which new members were trained, and did all the cutting and placing of stone at the building site. The basic tools of the medieval masons were little changed from those of Egypt, but they had large saws driven by waterwheels to cut stone as well as considerable machinery for raising and moving materials. Their knowledge of technics was a closely held secret; it included the rules of proportion for overall planning and for determining the safe dimensions of structural members. One extant sketchbook of drawings, from the master mason Villard de Honnecourt, shows a keen sense of observation, a love of mechanical devices, and above all the notion of geometric form that underlay the work; but it gives only tantalizing bits of information about actual building construction. Jean Mignot, one of the master masons of Milan Cathedral, summed up their approach with the phrase *ars sine scientia nihil est,* "art without science is nothing"; that is, skill in building derived from practical experience (*ars*) must be tempered and guided by precise principles (*scientia*), which were seen as being embodied in the theorems of geometry, the only science of medieval times. But with these limited means the masons were able to realize great achievements.

The "cathedral crusade"

Romanesque masons had two patrons, church and state. The state built mostly for military purposes, and Roman stonework, once recovered, was adequate for castles and fortifications. But the church had other interests that propelled the development of stone construction in new and daring directions. St. Augustine had written that light is the most direct manifestation of God. It was this idea that led the search for ways to introduce more and more light into churches, opening ever larger windows in the walls until a new kind of diaphanous stone skeleton evolved.

The Roman-inspired circular cross vaults and arches in

The Gothic arch

stone were heavy and needed heavy walls and piers to receive their thrusts; the windows they offered were small. Medieval masons found that there was a more efficient form for the arch than the classical circle; this form is a catenary curve—that is, one formed by a chain when it hangs under its own weight. But the masons' belief in geometry and the perfection of circular forms led them to approximate the catenary shape with two circular segments that met in a point at the top, the so-called Gothic arch. Such arches could be made thinner since they more efficiently channeled the compressive forces that flowed through them and allowed larger openings in the walls. The heavy piers that took the lateral thrust of the roof vaults were soon hollowed out into half arches or flying buttresses, which allowed even more light to enter the nave. To absorb the forces flowing down through the stone frame, massive foundations were required; often the volume of stone below ground was greater than that above. To further lighten the loads, the vaults themselves were made thinner by introducing ribs at the intersections of their curved surfaces, called groins. The ribs were built with supporting formwork or centring made of timber; close cooperation was needed between the carpenters and the masons. The curved surfaces of stones between the ribs were probably laid with little formwork, using only mortar; brick vaults are still built this way in the Middle East. The mortar was used not only for adhesion as a construction device but also later to check for tension cracks, which were signs of possible failure; the mortar thus served as a means of quality control to help keep the structure in compression. The naves of cathedrals were made higher to gather more light; Amiens Cathedral (begun 1220) was 42 metres (140 feet) high (Figure 5), and finally in 1347 Beauvais Cathedral reached the maximum height of 48 metres (157 feet), but its vaults soon collapsed and had to be rebuilt. The spans of the naves of Gothic churches remained fairly small, about 13 to 16 metres (45 to 55 feet); only a few late examples have longer spans, the greatest being 23 metres (74 feet) at Gerona Cathedral (completed 1458).

After the enthusiasm of the cathedral crusade ebbed in the 14th century and the basic fabric of most cathedrals was completed, a new element appeared to further test the skill of masons and carpenters: the spire. The spire was more a symbol of local pride than a part of the theological quest for more light, but it raised interesting technical problems. At Salisbury Cathedral the spire was built over the crossing of the nave and transept, which had not been designed to accommodate it; the tall crossing piers began to buckle under the added weight. Strainer arches had to be added between the piers to brace them against buckling; this was apparently the first time that stone columns were slender and heavily loaded enough to be observed to bend or buckle—later, such action would be a major concern in the design of metal columns. Salisbury's spire is an ingenious composite structure of stone cladding laid over a timber frame and tied together at the base with iron bands to resist spreading; it rose to a total height of 123 metres (404 feet) when it was finished in 1362. Strasbourg Cathedral added a 144-metre (475-foot) spire in 1439, and the upper limit was reached at Beauvais Cathedral in 1569 when its 157-metre (516-foot) spire was completed; the Beauvais spire collapsed in 1573 and was never rebuilt, a last sad epilogue to the cathedral crusade.

Half-timber construction

Construction in timber and brick. Timber construction underwent slow development in this period. Scandinavian stave churches of heavy timber were built in the 8th to 10th centuries, prior to the triumph of the stone church, and a few have survived to the present day. In western Europe, particularly from the 14th century onward, half-timber construction emerged as a new form of house building. The continental type had a frame of squared timbers, with vertical posts spaced about one metre apart and horizontal girts spaced at the same distance; diagonal braces were run through the outside walls for lateral stability. The roof beams spanned between the ridgepole and the walls; floor beams were supported on the walls and interior partitions. The English half-timber frame was similar, but it eliminated the horizontal girts and diagonal bracing by using closely spaced verticals about one-half metre apart. In both systems the space in the outside wall was filled with an enclosure material to impart added rigidity to the frame; brick or wattle and daub were often used. All the timbers of the frame were attached together by elaborate dovetail, or mortise-and-tenon, joints. Half-timber framing would remain the standard way of building with wood in Europe until the 19th century. There was also considerable use of heavy timber for the roofs and floors of masonry buildings, which was influenced by shipbuilding technology. A particular instance of this is the English hammer-beam roof, which was a kind of corbeled truss that could span quite long distances. The roof of King Richard II's Westminster Hall in London (1402), with a 21-metre (70-foot) span, is an excellent example of this type.

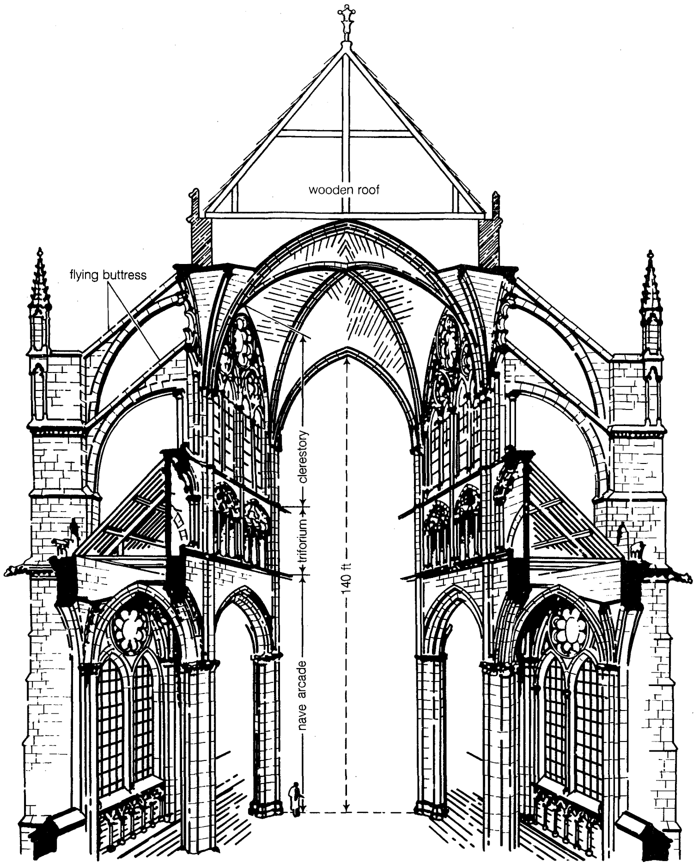

Figure 5: Transverse section of Amiens, a typical Gothic cathedral (1220–88).

The British Architectural Library, RIBA, London

Fired brick began to be made again in Europe in the 14th century, preceded in many areas by the use of salvaged Roman brick. The 14th-century bricks were not as precise as the Roman and were often distorted in firing. Therefore, large lime-mortar joints were needed for regular course lines. Bricks became nearly standardized at something close to the present size, about 20.3 × 9.5 × 5.7 centimetres (8 × 3.75 × 2.25 inches), and bonding systems based on this approximately 2 : 1 proportion were developed. These bonding patterns reduced continuous vertical mortar joints, because the mortars were of substantially lower strength than the bricks and vertical joints could form planes of weakness in the walls where cracks might develop. The best bonding pattern was English bond, in which all the bricks in each course overlapped the ones below and vertical joints were entirely eliminated. Brick remained quite expensive because of the cost of the fuel needed to fire it, and it was used mainly where there was no readily available stone. In the late medieval period and mostly in northern Europe, brick was adapted to Gothic stone forms to build so-called hall churches, with naves and aisles of equal height.

Building services. Although Roman hypocaust heating disappeared with the empire, a new development in interior heating appeared in western Europe at the beginning of the 12th century: the masonry fireplace and chimney began to replace the central open fire. The large roof openings over central fires let in wind and rain, so each house had only one and larger buildings had as few as possible. Therefore, heated rooms tended to be large and semipublic, where many persons could share the fire's warmth; the roof opening did not effectively remove all the smoke, some of which remained to plague the room's occupants. The chimney did not let in much air or water and could remove most of the smoke. Although much of the heat went up the flue, it was still a great improvement, and, most significantly, it could be used to heat both small and large rooms and multistory buildings as well. Houses, particularly large ones, were broken up into smaller, more private spaces each heated by its own fireplace, a change that decisively altered the communal life-style of early medieval times.

THE RENAISSANCE

Reintroduction of dome construction. The waning of the cathedral crusade in the late 14th century led to a decline in the International Gothic style practiced by the master masons. In this period the emerging nation-states of Europe began to compete with the church as centres of power. To these new nations, the Roman Empire was the model nation-state, and it seemed appropriate that they use Roman building forms as symbols of their power—particularly the round arch, the vault, and, above all, the dome, following the powerful example of the Pantheon. From 1350 until 1750 much of building technology was focused on the domed church, which developed as a symbol not only of religious belief but also of national and urban pride. There was a conscious rejection of Gothic forms in favour of the ideological appeal of Rome. This attitude led to a split between the processes of design and construction and to the appearance of the first architects (a word derived from the Greek *architekton,* meaning a chief craftsman), who conceived a building's form, as opposed to the builder, who executed it. The first building in which the designer and the builder were separate persons was the Campanile, or bell tower, of the cathedral of Florence. The design was made by the painter Giotto and constructed by cathedral masons from 1334 to 1359.

The Duomo

The cathedral of Florence itself had been begun in the Gothic style by Arnolfo di Cambio in 1296. But in 1366 the City of Florence, following the advice of certain painters and sculptors, decided that the Gothic should no longer be used and that all new work should follow Roman forms, including an octagonal dome 42 metres (138 feet) in span to be built at the east end of the nave. The dome was not built until the early 15th century, when Filippo Brunelleschi, a goldsmith and sculptor, began to make statues for the cathedral. Gradually he became interested in the building itself and built some smaller parts of it. In about 1415 he prepared a design for the dome that he daringly proposed to build without the aid of formwork, which had been absolutely necessary in all previous Roman and Gothic construction. He built a 1 : 12 model of the dome in brick to demonstrate his method; the design was accepted and built under his supervision from 1420 to 1436. Brunelleschi was thus the first real architect to conceive the building's form and the methods to execute it and to guarantee its performance; he pointedly refused membership in both the masons' and carpenters' guilds. Brunelleschi's dome (Figure 6) consists of two layers, an inner dome spanning the diameter and a parallel outer shell to protect it from the weather and give it a more pleasing external form. Both domes are supported by 24 stone half arches, or ribs, of circular form, 2.1 metres (seven feet) thick at the base and tapering to 1.5 metres (five feet), which meet at an open stone compression ring at the top. To resist outward thrust, tie rings of stone held together with metal cramps run horizontally between the ribs. There are also tie rings of oak timbers joined by metal connectors. The spaces between the ribs and tie rings are spanned by the inner and outer shells, which are of stone for the first 7.1 metres (23 feet) and brick above. The entire structure was built without formwork, the circular profiles of the ribs and rings being maintained by a system of measuring wires fixed at the centres of curvature. Brunelleschi obviously understood enough about the structural behaviour of the dome to know that, if it were built in horizontal layers, it would always be stable and not require timber centring. He also designed elaborate wooden machines to move the needed building materials both vertically and horizontally. Having all but equaled the span of the Pantheon in stone, Brunelleschi was hailed as the man who "renewed Roman masonry work"; the dome was established as the paragon of built form.

The next great dome of the Renaissance was that of St. Peter's Basilica in Rome, begun by Pope Julius II in 1506. The technology was very similar to that of Brunelleschi, and the diameter is nearly the same. The dome's design went through many changes and extended over a period of nearly 80 years. The major contributors to the design were the painter and sculptor Michelangelo, who served as architect from 1546 to 1564, and the architects Giacomo della Porta and Domenico Fontana, under whose direction it was finally built during the 1580s. The dome was considerably thinner than that of Florence and was reinforced by three tie rings made of continuous iron chains. It developed numerous cracks, and in the 1740s five more chains were added to further stabilize it. Since the dome used a proven technology, most of the design was done on paper with drawings.

St. Paul's Cathedral

Another large dome of this period was that of St. Paul's Cathedral in London, which was built from 1675 to 1710 by the English architect Sir Christopher Wren. In the early stages of the design process only two physical models were used; later efforts included extensive drawings and apparently also mathematical modeling with numerical calculations. Wren had begun his career as a mathematician and physical scientist and was professor of astronomy at Oxford from 1661 to 1673 before becoming a full-time architect. With this background he was thus able to profit from the first theoretical determination of the catenary curve as the most efficient profile of the arch and dome, which was published by the Scottish mathematician David Gregory in 1697. Wren's solution to the dome, which has a diameter of 34.5 metres (113 feet), was a series of three nested shells, of which the middle one is the true structure. This middle dome is built of brick in a nearly

Adapted from F.D. Prager and G. Scaglia, *Brunelleschi: Studies of His Technology and Inventions* (1970), Massachusetts Institute of Technology Press

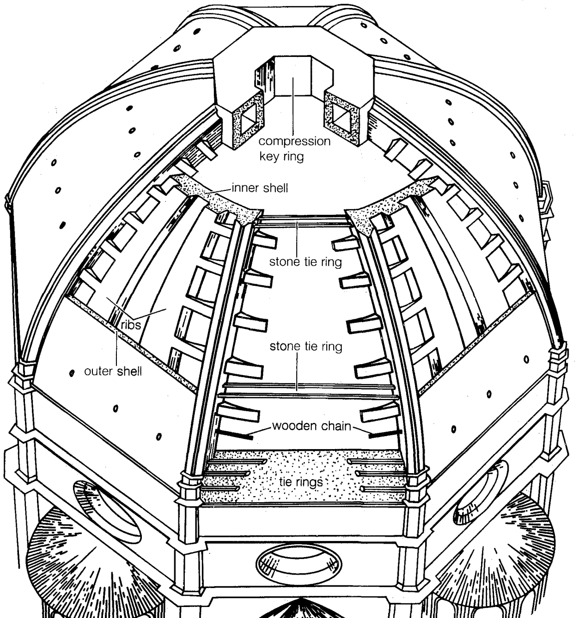

Figure 6: Dome construction of the Cathedral of Santa Maria del Fiore, Florence; Filippo Brunelleschi, architect (1420–36).

conical catenary form, due to the large concentrated load of the lantern on top, and constrained by iron chains; it supports a triangularly braced timber framework to which is attached the exterior surfacing of lead sheets. Within the middle dome is a shallower catenary dome that carries only its own weight and serves as a ceiling for the interior space. Wren's concealed structure, to which were applied the desired internal and external forms, has become a standard architectural technique.

Revival of Roman technics and materials. In addition to Roman forms in masonry, the Renaissance recovered other Roman technologies, including timber trusses. Giorgio Vasari used king-post timber trusses for a 20-metre (66-foot) span in the roof of the Uffizi, or Municipal Office Building, in Florence in the mid-16th century. At the same time, the Venetian architect Andrea Palladio used a fully triangulated timber truss for a bridge with a span of 30.5 metres (100 feet) over the Cimone River. Palladio clearly understood the importance of the carefully detailed diagonal members, for in his diagram of the truss in his *Four Books on Architecture* he said that they "support the whole work." The tension connections of the timber members in the truss were joined with iron cramps and bolts.

Trussed spans in the range 20–26 metres (65–85 feet) became fairly common in building roofs. In 1664 Wren used timber trusses with a span of about 22 metres (73 feet) in the roof of the Sheldonian Theatre at Oxford. But a precise theoretical understanding of the truss, and major use of it in buildings, would not come until the 19th century.

Reintroduction of clear glazing

Another Roman material that was revived and much improved in the Renaissance was clear glass. A new technique for making it was perfected in Venice in the 16th century. It was known as the crown glass method and was originally used for making dinner plates. Glassblowers spun the molten glass into flat disks up to a metre in diameter; the disks were polished after they had cooled and were cut into rectangular shapes. The first record of crown glass windows is their installation in double-hung counterweighted sliding-sash frames, at Inigo Jones's Banqueting House in London in 1685. Large areas of such glass became common in the 1700s, pointing the way toward the great glass and iron buildings of the 19th century.

The efficiency of interior heating was improved by the introduction of cast-iron and clay-tile stoves, which were placed in a free-standing position in the room. The radiant heat they produced was uniformly distributed in the space, and they lent themselves to the burning of coal—a new fuel that was rapidly replacing wood in western Europe. When European builders had recovered the technology of the classical world in brick, stone, and timber, a stable plateau was reached in the development of the building arts; these materials and technics were well-suited to the churches, palaces, and fortifications that their patrons required. The Industrial Revolution, however, brought new materials and the demand for new building types that completely transformed building technology.

THE FIRST INDUSTRIAL AGE

Development of iron technology. The last half of the 18th century saw the unfolding of a series of events, primarily in England, that later historians would call the first Industrial Revolution, which would have a profound influence on society as a whole as well as on building technology. Among the first of these events was the large-scale production of iron, beginning with the work of Abraham Darby, who in 1709 was the first to use coke as a fuel in the smelting process. The ready availability of iron contributed to the development of machinery, notably James Watt's double-acting steam engine of 1769. Henry Cort developed the puddling process for making wrought iron in 1784, and in the same year he built the first rolling mill, powered by a steam engine, to produce rolled lengths of wrought-iron bars, angles, and other shapes. Cast iron, which has a higher carbon content than wrought iron but is more brittle, was also produced on a large scale. Standard iron building elements soon appeared, pointing the way to the development of metal buildings.

Early applications of iron in construction are found several centuries prior to the industrial age. There are records of iron chain suspension bridges with timber decks in China from the early Ming dynasty (1368–1644); some of them, such as the Liu-Tung Bridge, the object of a famous battle on Mao Zedong's (Mao Tse-tung's) Long March in 1935, have survived in a much-restored condition. The iron tension chains in the domes of St. Peter's and St. Paul's cathedrals are other examples. But the first large cast-iron structure of the industrial age was the bridge over the River Severn at Coalbrookdale. Built by the iron founders Abraham Darby III and John Wilkinson between 1775 and 1779, it has a span of 30 metres (100 feet), using five circular-form arches that are reduced to a spidery web of slender iron ribs. Each arch was cast in two pieces with a maximum dimension of 21 metres (70 feet), which were difficult to move from the foundry to the site and to set in place. Smaller, more easily handled pieces characterized the rapid application of iron to buildings that followed. Solid cast-iron columns were used in St. Anne's Church in Liverpool as early as 1772, and hollow tubular columns of increased efficiency were developed in the 1790s. The first use of wrought-iron trusses, which were made of flat bars riveted together, was in a 28-metre (92-foot) span for the roof of the Théâtre-Français in Paris in 1786 by the architect Victor Louis; there iron was used not so much for its strength as its noncombustibility, which, it was hoped, would reduce the hazard of fire. For the same reason, in about 1800 the British textile industry began to use partial metal framing in mill buildings up to seven stories high. Hollow cast-iron cylindrical columns were spaced at about three metres (10 feet) on centre and supported cast-iron tee beams spanning up to 4.5 metres (15 feet); the floors were bridged by brick arches resting on the bottom flanges of the tee beams; at the perimeter the beams rested on masonry bearing walls, which gave the structure its lateral stability. This prototype of the iron-frame building with exterior masonry walls soon set a standard that would continue to the end of the century.

The iron bridge at Coalbrookdale

The completely independent iron frame without masonry adjuncts emerged slowly in a series of special building types. The first modest example was Hungerford Fish Market (1835) in London. Timber was forbidden because of sanitation requirements; the cast-iron beams spanned 9.7 metres (32 feet) with three-metre (10-foot) cantilevers on either side, and the hollow cast-iron columns also served as roof drains. All lateral stability was provided by the rigid joints between columns and beams. The next type to use the full iron frame was the greenhouse, which provided a controlled luminous and thermal environment for exotic tropical plants in the cold climate of northern Europe. Among the first of these was the Palm House at Kew Gardens near London; it was built by the architect Decimus Burton in the 1840s.

A spectacular series of iron and glass buildings for conservatories and exhibition halls continued to the end of the century. The most important of these was the Crystal Palace, built in London's Hyde Park to house the Great Exhibition of 1851. This vast building, 564 metres (1,851 feet) long, was built entirely of standardized parts. Cast-iron columns carried iron trusses of three different spans—7.3 metres (24 feet), 14.6 metres (48 feet), and 21.9 metres (72 feet)—in riveted wrought iron; spanning between the trusses were ingenious "Paxton gutters" made of wooden compression members above iron tension rods that prestressed the wood to reduce deflection. All these prefabricated elements were simply bolted or clipped together on the site to enclose a space of 90,000 square metres (1,000,000 square feet) in only six months. But the major triumph of the Crystal Palace was its all-glass enclosure, made of standard panes 25 × 124 centimetres (10 × 49 inches) in size; the huge space was flooded with light that was scarcely interrupted by the diaphanous metal framing—it resembled a great secular cathedral realizing the ultimate ambition of the medieval masons.

The French also produced a number of fine iron and glass exhibition halls, including one with a 48-metre (160-foot) span in 1855. Others with somewhat smaller spans, but larger enclosed areas than the Crystal Palace, followed in 1867 and 1878. Iron trusses with glazed roofs were also used in the train sheds of railway stations that were built throughout western Europe. The New Street Station in Birmingham, Eng. (1854), had a train shed with an iron truss roof spanning 64 metres (211 feet). It was apparently the first building to exceed the span of the Pantheon. One of the largest was St. Pancras Station (1873) in London, which featured a glazed hall spanned by 74-metre (243-foot) trussed iron arches. After the brilliant successes of midcentury, iron and glass construction was applied in a more prosaic series of buildings that continued to be built until 1900.

Manufactured building materials. The production of brick was industrialized in the 19th century. The laborious process of hand-molding, which had been used for 3,000 years, was superseded by "pressed" bricks. These were mass-produced by a mechanical extrusion process in which clay was squeezed through a rectangular die as a continuous column and sliced to size by a wire cutter. There was also a proliferation of elaborately shaped and stamped masonry units. Periodically fired beehive kilns (stoked by coke) continued to be used, but the continuous tunnel kiln, through which bricks were moved slowly on a conveyor belt, had appeared by the end of the century. The new methods considerably reduced the cost of brick, and it became one of the constituent building materials of the age.

The balloon frame

Timber technology underwent rapid development in the 19th century in North America, where there were large forests of softwood fir and pine trees that could be harvested and processed by industrial methods; steam- and water-powered sawmills began producing standard-dimension timbers in quantity in the 1820s. The production of cheap machine-made nails in the 1830s provided the other necessary ingredient that made possible a major innovation in building construction, the balloon frame; the first example is thought to be a warehouse erected in Chicago in 1832 by George W. Snow. There was a great demand for small buildings of all types as the North American continent was settled, and the light timber frame provided a quick, flexible, and inexpensive solution to this problem. In the balloon frame system, traditional heavy timbers and complex joinery were abandoned. The building walls were framed with 5 × 10-centimetre (2 × 4-inch) vertical members, or studs, placed at 40 centimetres (16 inches) on centre (that is, measured between the centre points of each); these in turn supported the roof and floor joists, usually 5 × 25 centimetres (2 × 10 inches) also placed 40 centimetres (16 inches) apart and capable of spanning up to six metres (20 feet). Lateral stability was achieved by light diagonal braces let into the studs or, more commonly, by two-centimetre (0.75-inch) thick diagonal boards applied to all exterior walls and to floor and roof joists, creating a rigid, light box. Openings were cut through the framing and sheathing as required. All connections were made with machine-made nails, which were easily driven through the soft, thin timbers. A wide variety of interior and exterior surfacing materials could be applied to the frame, including timber siding, stucco, and brick veneer. The balloon frame building, made with manufactured materials and requiring only a few hand tools and little skill to build, has remained a popular and inexpensive form of construction to the present day.

Building science. A significant achievement of the first industrial age was the emergence of building science, particularly the elastic theory of structures. With it, mathematical models could be used to predict structural performance with considerable accuracy, provided there was adequate quality control of the materials used. Although some elements of the elastic theory, such as the Swiss mathematician Leonhard Euler's theory of column buckling (1757), were worked out earlier, the real development began with the English scientist Thomas Young's modern definition of the modulus of elasticity in 1807. Louis Navier published the elastic theory of beams in 1826, and three methods of analyzing forces in trusses were devised by Squire Whipple, A. Ritter, and James Clerk Maxwell between 1847 and 1864. The concept of a statically determinate structure—that is, a structure whose forces could be determined from Newton's laws of motion alone—was set forth by Otto Mohr in 1874, after having been used intuitively for perhaps 40 years. Most 19th-century structures were purposely designed and fabricated with pin joints to be statically determinate; it was not until the 20th century that statically indeterminate structures became readily solvable. The elastic theory formed the basis of structural analysis until World War II, when bomb-damaged buildings were observed to behave in unpredicted ways and the underlying assumptions of the theory were found to require modification.

Emergence of design professionals. The coming of the industrial age also marked a major change in the role of the architect. The artist-architects of the Renaissance had the twin patrons of church and state upon whom they could depend for commissions. In the rising industrial democracies the market for large-scale buildings worthy of an architect's attention widened, and the different users asked for a bewildering range of new building types. The response of the architect was to develop the new role of licensed professional on the model of professions such as law and medicine. In addition, with the coming of building science, there was a further division of labour in the design process; structural engineering appeared as a separate discipline specializing in the application of mathematical models in building. One of the first buildings for which the architect and engineer were separate persons was the Granary (1811) in Paris. Societies representing the building design professions were founded, including the Institution of Civil Engineers (1818) and the Royal Institute of British Architects (1834), both in London, and the American Institute of Architects (1857). Official government licensing of architects and engineers, a goal of these societies, was not realized until much later, beginning with the Illinois Architects Act of 1897. Concurrent with the rise of professionalism was the development of government regulation, which took the form of detailed municipal and national building codes specifying both prescriptive and performance requirements for buildings.

Gas lighting

Improvements in building services. Environmental control technologies began to develop dramatically in the first industrial age. The first major advance was the use of coal gas for lighting. Coal gas was first made in the 1690s by heating coal in the presence of water to yield methane, and in 1792 William Murdock developed the gas jet lighting fixture. The first large building to have gas lighting (from a small gas plant on the site) was James Watt's foundry in Birmingham in 1803. The Gas Light and Coke Company was founded in London in 1812 as the first real public utility, producing coal gas as a part of the coking process in large central plants and distributing it through underground pipes to individual users; soon many major cities had gasworks and distribution networks. Gas was expensive, however, and was used mainly for lighting, not for heating or cooking; it also contained many impurities that produced undesirable products of combustion (particularly carbon soot) in occupied spaces. Relatively pure methane in the form of natural gas would not be available until the exploitation of large oil fields in the 20th century.

The stove and fireplace continued as the major sources of space heating throughout this period, but the development of the steam engine and its associated boilers led to a new technology in the form of steam heating. James Watt heated his own office with steam running through pipes as early as 1784. During the 19th century, systems of steam and later hot-water heating were gradually developed; these used coal-fired central boilers connected to networks of pipes that distributed the heated fluid to cast-iron radiators and returned it to the boiler for reheating. Steam heat was a major improvement over stoves and fireplaces because all combustion products were eliminated from occupied spaces, but heat sources were still localized at the radiators.

Plumbing and sanitation systems in buildings advanced rapidly in this period. Public water-distribution systems were the essential element; the first large-scale example of a mechanically pressurized water-supply system was the great array of waterwheels installed by Louis XIV at Marley on the Marne River in France to pump water for the fountains at Versailles, about 18 kilometres (10 miles)

Advances in plumbing

away. The widespread use of cast-iron pipes in the late 18th century made higher pressures possible, and they were used by Napoleon in the first steam-powered municipal water supply for a section of Paris in 1812. Gravity-powered underground drainage systems were installed along with water-distribution networks in most large cities of the industrial world during the 19th century; sewage-treatment plants were introduced in the 1860s. Permanent plumbing fixtures appeared in buildings with water supply and drainage, replacing portable basins, buckets, and chamber pots. Joseph Bramah invented the metal valve-type water closet as early as 1778, and other early lavatories, sinks, and bathtubs were of metal also; lead, copper, and zinc were all tried. The metal fixtures proved difficult to clean, however, and in England during the 1870s Thomas Twyford developed the first large one-piece ceramic lavatories as well as the ceramic washdown water closet. At first these ceramic fixtures were very expensive, but their prices declined until they became standard, and their forms remain largely unchanged today. The bathtub proved to be too large for brittle ceramic construction, and the porcelain-enamel cast-iron tub was devised about 1870; the double-shell built-in type still common today appeared about 1915.

THE SECOND INDUSTRIAL AGE

Introduction of steel building technology. If the first industrial age was one of iron and steam, the second industrial age, which began in about 1880, could be called one of steel and electricity. Mass production of this new material and of this new form of energy also transformed building technology. Steel was first made in large quantities for railroad rails. Rolling of steel rails (which was adapted from wrought-iron rolling technology) and other shapes such as angles and channels began about 1870; it made a much tougher, less brittle metal. Steel was chosen as the principal building material for two structures built for the Paris Exposition of 1889: the Eiffel Tower and the Gallery of Machines. Gustave Eiffel's tower was 300 metres (1,000 feet) high, and its familiar parabolic curved form has become a symbol of Paris itself; its height was not exceeded until 1929. The Gallery of Machines (Figure 7) was designed by the architect C.-L.-F. Dutert and the engineer Victor Contamin with great three-hinged arches spanning 114 metres (380 feet) and extending more than 420 metres (1,400 feet). Its glass-enclosed clear span area of 48,727 square metres (536,000 square feet) has never been equaled; in fact, it was so large that no regular use for it could be found after the exposition closed, and this magnificent building was demolished in 1910.

Early high-rise construction in steel. While these prodigious structures were the centre of attention, a new and more significant technology was developing: the steel-framed high-rise building. It began in Chicago, a city whose central business district was growing rapidly. The pressure of land values in the early 1880s led owners to demand taller buildings. The architect-engineer William Le Baron Jenney responded to this challenge with the 10-story Home Insurance Company Building (1885), which had a nearly completely all-metal structure. The frame consisted of cast-iron columns supporting wrought-iron beams, together with two floors of rolled-steel beams that were substituted during construction; this was the first large-scale use of steel in a building. The metal framing was completely encased in brick or clay-tile cladding for fire protection, since iron and steel begin to lose strength if they are heated above about 400° C (750° F). Jenney's Manhattan Building (1891) had the first vertical truss bracing to resist wind forces; rigid frame or portal wind bracing was first used in the neighbouring Old Colony Building (1893) by the architects William Holabird and Martin Roche. The all-steel frame finally appeared in Jenney's Ludington Building (1891) and the Fair Store (1892; Figure 8).

The Home Insurance Company Building

The foundations of these high-rise buildings posed a major problem, given the soft clay soil of central Chicago. Traditional spread footings, which dated back to the Egyptians, proved to be inadequate to resist settlement due to the heavy loads of the many floors, and timber piles (a Roman invention) were driven down to bedrock. For the 13-story Stock Exchange Building (1892), the engineer Dankmar Adler employed the caisson foundation used in bridge construction. A cylindrical shaft braced with board sheathing was hand-dug to bedrock and filled with concrete to create a solid pier to receive the heavy loads of the steel columns.

By 1895 a mature high-rise building technology had been developed: the frame of rolled steel I beams with bolted or riveted connections, diagonal or portal wind bracing, clay-tile fireproofing, and caisson foundations. The electric-powered elevator provided vertical transportation, but other environmental technologies were still fairly simple. Interior lighting was still largely from daylight, although supplemented by electric light. There was steam heating but no cooling, and ventilation was dependent on operating windows; thus these buildings needed narrow floor spaces to give adequate access to light and air. Of equal importance in high-rise construction was the introduction of the internal-combustion engine (which had been invented by Nikolaus Otto in 1876) at the building site; it replaced the horse and human muscle power for the heaviest tasks of lifting. Over the next 35 years, higher steel-frame buildings were built; in Chicago the Masonic Building (1892) of Daniel Burnham and John Root reached 21 stories (91 metres or 302 feet), but then the leadership shifted to New York City with the 26-story St. Paul Building (1899). The Singer Building (1907) by the architect Ernest Flagg rose to 47 stories (184 metres or 612 feet), Cass

Support systems in high-rise buildings

Historical Picture Service, Chicago

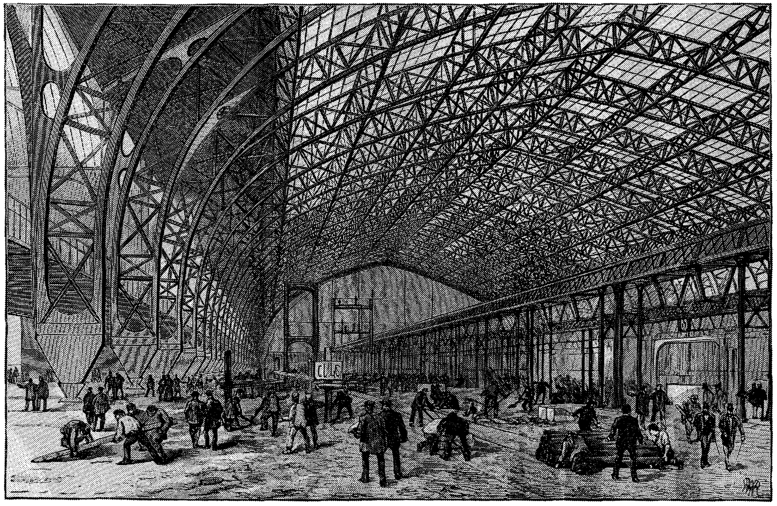

Figure 7: Steel and glass construction using three-hinge arches, the Gallery of Machines, Paris, 1889 (demolished 1910).

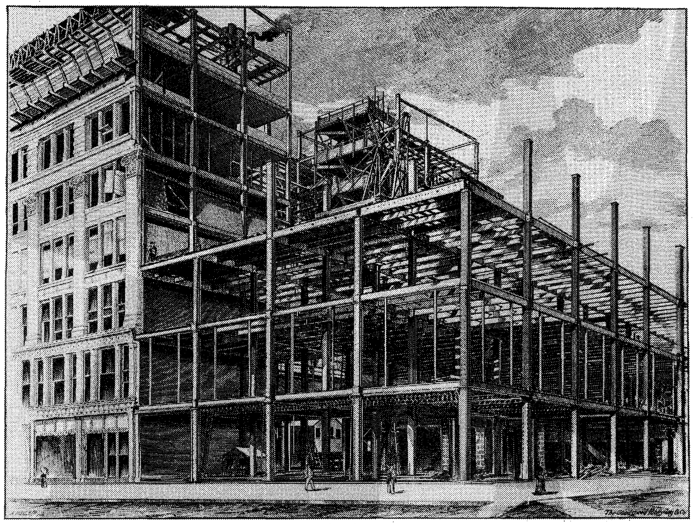

Figure 8: Structural elements of a steel-frame high-rise building under construction, the Fair Store, Chicago, 1892 (demolished 1984), from *Industrial Chicago,* 1891.

Chicago Historical Society, ICHi 21294

Gilbert's Woolworth Building (1913) attained a height of 238 metres (792 feet) at 55 stories, and Shreve, Lamb & Harmon's 102-story Empire State Building (1931) touched 378 metres (1,239 feet). The race for higher buildings came to an abrupt halt with the Great Depression and World War II, and high-rise construction was not resumed until the late 1940s.

Long-span construction in steel. Long-span structures in steel developed more slowly than the high-rise in the years from 1895 to 1945, and none exceeded the span of the Gallery of Machines. Two-hinge (made of a single member hinged at each end) and three-hinge (made of two members hinged at each end and at the meeting point at the crown) trussed arches were widely used, the largest examples being two great airship hangars for the U.S. Navy in New Jersey—the first built in 1922 with a span of 79 metres (262 feet), the second in 1942 with a span of 100 metres (328 feet). The flat truss was used also, reaching a maximum span of 91 metres (300 feet) in the Glenn L. Martin Co. Aircraft Assembly Building (1937) in Baltimore. Electric arc welding, another important steel technology, was applied to building construction at this time, although the principle had been developed in the 1880s. The first-all-welded multistory buildings were a series of factories for the Westinghouse Company, beginning in 1920. The welded rigid frame became a new structural type for medium spans, reaching a length of 23 metres (77 feet) in the Cincinnati Union Terminal (1932), but widespread use of welding did not come until after 1945.

Reintroduction of concrete. The second industrial age also saw the reemergence of concrete in a new composite relationship with steel, creating a technology that would rapidly assume a major role in building construction. The first step in this process was the creation of higher-strength artificial cements. Lime mortar—made of lime, sand, and water—had been known since ancient times. It was improved in the late 18th century by the British engineer John Smeaton, who added powdered brick to the mix and made the first modern concrete by adding pebbles as coarse aggregate. Joseph Aspdin patented the first true artificial cement, which he called Portland Cement, in 1824; the name implied that it was of the same high quality as Portland stone. To make portland cement, Aspdin burned limestone and clay together in a kiln; the clay provided silicon compounds, which when combined with water formed stronger bonds than the calcium compounds of limestone. In the 1830s Charles Johnson, another British cement manufacturer, saw the importance of high-temperature burning of the clay and limestone to a white heat, at which point they begin to fuse. In this period, plain concrete was used for walls, and it sometimes replaced brick in floor arches that spanned between wrought-iron beams in iron-framed factories. Precast concrete blocks also were manufactured, although they did not effectively compete with brick until the 20th century.

Portland cement

The invention of reinforced concrete. The first use of iron-reinforced concrete was by the French builder François Coignet in Paris in the 1850s. Coignet's own all-concrete house in Paris (1862), the roofs and floors reinforced with small wrought-iron I beams, still stands. But reinforced concrete development began with the French gardener Joseph Monier's 1867 patent for large concrete flowerpots reinforced with a cage of iron wires. The French builder François Hennebique applied Monier's ideas to floors, using iron rods to reinforce concrete beams and slabs; Hennebique was the first to realize that the rods had to be bent upward to take negative moment near supports. In 1892 he closed his construction business and became a consulting engineer, building many structures with concrete frames composed of columns, beams, and slabs. In the United States Ernest Ransome paralleled Hennebique's work, constructing factory buildings in concrete. High-rise structures in concrete followed the paradigm of the steel frame. Examples include the 16-story Ingalls Building (1903) in Cincinnati, which was 54 metres (180 feet) tall and the 11-story Royal Liver Building (1909), built in Liverpool by Hennebique's English representative, Louis Mouchel. The latter was Europe's first skyscraper, its clock tower reaching a height of 95 metres (316 feet). Attainment of height in concrete buildings progressed slowly due to the much lower strength and stiffness of concrete as compared to steel.

Between 1900 and 1910 the elastic theory of structures was at last applied to reinforced concrete in a scientific way. Emil Morsch, the chief engineer of the German firm of Wayss and Freitag, formulated the theory, which was verified by detailed experimental testing at the Technical University of Stuttgart. These tests established the need for deformed bars for good bonding with concrete and demonstrated that the amount of steel in any member should be limited to about 8 percent of the area; this assures the slow elastic failure of the steel, as opposed to the abrupt brittle failure of the concrete, in case of accidental overloading. In 1930 the American engineer Hardy Cross introduced relaxation methods for the approximate analysis of rigid frames, which greatly simplified the design of concrete structures. In the Johnson-Bovey Building (1905) in Minneapolis, the American engineer C.A.P. Turner employed concrete floor slabs without beams (called flat slabs or flat plates) that used diagonal and orthogonal patterns of reinforcing bars. The system still used today—which divides the bays between columns into column strips and middle strips and uses only an orthogonal arrangement of bars—was devised in 1912 by the Swiss engineer Robert Maillart.

The concrete dome. Concrete was also applied to long-span buildings, an early example being the Centennial Hall (1913) at Breslau, Ger. (now Wrocław, Pol.), by the architect Max Berg and the engineers Dyckerhoff & Widmann; its ribbed dome spanned 65 metres (216 feet), exceeding the span of the Pantheon. More spectacular were the great airship hangars at Orly constructed by the French engineer Eugène Freyssinet in 1916; they were made with nine-centimetre- (3.5-inch-) thick corrugated parabolic vaults spanning 80 metres (266 feet) and pierced by windows. In the 1920s Freyssinet made a major contribution to concrete technology with the introduction of pretensioning. In this process, the reinforcing wires were stretched in tension, and the concrete was poured around them; when the concrete hardened, the wires were released, and the member acquired an upward deflection and was entirely in compression. When the service load was applied, the member deflected downward to a flat position, remaining entirely in compression, and it did not develop the tension cracks that plague ordinary reinforced concrete. Widespread application of pretensioning was not made until after 1945.

Shell construction in concrete also began in the 1920s; the first example was a very thin (6 centimetres) hemispherical shell for a planetarium (1924) in Jena, Ger., spanning 25 metres (82 feet). In 1927 an octagonal ribbed shell dome with a span of 66 metres (220 feet) was built to house a market hall in Leipzig. Many variations of thin shells were devised for use in industrial buildings. The shell emerged as a major form of long-span concrete structure after World War II.

Development of building service and support systems. *Vertical transportation.* Elisha Graves Otis developed the first safe steam-powered roped elevators with toothed guide rails and catches in the late 1850s. The steam-powered hydraulic elevator, which was limited to buildings of about 15 stories, was developed in 1867 by the French engineer Léon Édoux. The development of the electric motor by George Westinghouse in 1887 made possible the invention of the high-speed electric-powered roped elevator (called "lightning" elevators in comparison to the slower hydraulics) in 1889 and the electric-powered moving staircase, or escalator, in the 1890s.

Lighting. In the second industrial age, environmental technologies developed rapidly. Most of these technologies involved the use of electric power, which declined in cost during this period. The carbon-arc electric light was demonstrated as early as 1808, and the British physicist Michael Faraday devised the first steam-powered electric generator to operate a large carbon-arc lamp for the Foreland Lighthouse in 1858. But the carbon-arc lamp was so bright and required so much power that it was never widely used and was rapidly superseded by the simultaneous invention of the carbon-filament bulb by Thomas Edison and Joseph Swan in 1879. The carbon-filament bulb was highly inefficient, but it banished the soot and fire hazards of coal-gas jets and soon gained wide acceptance. It was succeeded by the more efficient tungsten-filament incandescent bulb, developed by George Coolidge of the General Electric Company, which first appeared in 1908; the double-coiled filament used today was introduced about 1930.

Carbon-arc light

Edison experimented with gas-discharge light tubes in 1896, and Georges Claude in France and Moore in England produced the first practical discharge tubes using noble gases such as neon and argon; these tubes were first used to outline the facade of the West End Cinema in London in 1913 and were rapidly exploited for signs and other decorative purposes. In 1938 General Electric and Westinghouse produced the first commercial fluorescent discharge lamps using mercury vapour and phosphor-coated tubes to enhance visible light output. Fluorescent tubes had roughly double the efficiency of tungsten lamps and were rapidly adopted for commercial and office use. Light intensity increased in all buildings as electric costs decreased, reaching a peak in about 1970. Gaseous-discharge lamps using high-pressure mercury and sodium vapour were developed in the 1960s but found only limited application in buildings; they are of such high intensity and marked colour that they are used mostly in high-ceilinged spaces and for exterior lighting.

Heating and cooling systems. Steam and hot-water heating systems of the late 19th century provided a reasonable means for winter heating, but no practical methods existed for artificial cooling, ventilating, or humidity control. In the forced-air system of heating, air replaced steam or water as the fluid medium of heat transfer, but this was dependent on the development of powered fans to move the air. Although large, crude fans for industrial applications in the ventilation of ships and mines had appeared by the 1860s, and the Johns Hopkins Hospital in Baltimore had a successful steam-powered forced-air system installed in 1873, the widespread application of this system to buildings only followed the development of electric-powered fans in the 1890s.

Important innovations in cooling technology followed. The development of refrigeration machines for food storage played a role, but the key element was Willis Carrier's 1906 patent that solved the problem of humidity removal by condensing the water vapour on droplets of cold water sprayed into an airstream. Starting with humidity control in tobacco and textile factories, Carrier slowly developed his system of "man-made weather," finally applying it together with heating, cooling, and control devices as a complete system in Graumann's Metropolitan Theater, Los Angeles, in 1922. The first office building air-conditioned by Carrier was the 21-story Milam Building (1928) in San Antonio, Texas. It had a central refrigeration plant in the basement that supplied cold water to small air-handling units on every other floor; these supplied conditioned air to each office space through ducts in the ceiling; the air was returned through grills in doors to the corridors and then back to the air-handling units. A somewhat different system was adopted by Carrier for the 32-story Philadelphia Savings Fund Society Building (1932). The central air-handling units were placed with the refrigeration plant on the 20th floor, and conditioned air was distributed through vertical ducts to the occupied floors and horizontally to each room and returned through the corridors to vertical exhaust ducts that carried it back to the central plant. Both systems of air handling, local and central, are still used in high-rise buildings. The Great Depression and World War II reduced the demand for air-conditioning systems, and it was not until the building of the United Nations Secretariat in New York City in 1949 that Carrier produced a method of air conditioning that could deal effectively with the large heat loads imposed by the building's all-glass curtain walls. The conditioned air was delivered not only from the ceiling but also through pipe coil convector units just inside the glass wall. The pipe coil convectors contained centrally supplied warm or cold water to further temper the heat loss or gain at the perimeter; conditioned air and water were centrally supplied from four mechanical floors spaced within the building's 39-story height.

Carrier's "man-made weather"

Carrier's "Weathermaster" system was energy-intensive, appropriate to the declining energy costs of the time, and it was adopted for most of the all-glass skyscrapers that followed in the next 25 years. In the 1960s the so-called dual-duct system appeared; both warm and cold air were centrally supplied to every part of the building and combined in mixing boxes to provide the appropriate atmosphere. The dual-duct system also consumed much energy, and, when energy prices began to rise in the 1970s, both it and the Weathermaster system were supplanted by the variable air volume (VAV) system, which supplies conditioned air at a single temperature, the volume varying according to the heat loss or gain in the occupied spaces. The VAV system requires much less energy and is widely used.

In the early 1950s, air-conditioning systems were reduced to very small electric-powered units capable of cooling single rooms. These were usually mounted in windows to take in fresh air and to remove heat to the atmosphere. These units found widespread application in the retrofitting of existing buildings—particularly houses and apartment buildings—and have since found considerable application in new residential buildings.

The relatively high energy costs of the 1970s also prompted interest in various forms of solar heating, both for interior spaces and for domestic hot water, but, except for residential passive solar heating, the relative decline in energy prices in the 1980s made such systems unattractive.

Use of thermal insulation

The study of thermodynamics in the late 19th century included the heat-transfer properties of materials and led to the concept of thermal insulation—that is, a material that has a relatively low rate of heat transfer. As building atmospheres became more carefully controlled after 1900, more attention was given to the thermal insulation of building enclosures (envelopes). One of the best insulators is air, and materials that trap air in small units have low heat-transfer rates; wool and foam are excellent examples. The first commercial insulations, in the 1920s, were mineral wools and vegetable-fibreboards; fibreglass wool appeared in 1938. Foam glass, the first rigid insulating foam, was marketed in the 1930s, and after 1945 a wide variety of plastic foam insulations was developed. Since the 1970s most building codes have set minimum requirements for insulation of building envelopes, and these have proved to be very cost-effective in saving energy.

Glass as a building material. Glass underwent considerable development in the second industrial age. The making of clear plate glass was perfected in the late 19th century, as were techniques of sandblasting and etching it. In the United States in 1905 the Libbey Owens Glass Company began making sheet glass by a continuous drawing process from a reservoir of molten glass; its surface was somewhat distorted, but it was much cheaper than plate glass. Prefabricated panels of double glazing about 2.5 centimetres (one inch) thick were first made in the 1940s, although the insulating principle of air trapped between two layers of glass had been recognized much earlier. Hollow glass blocks were introduced by the Corning Company in 1935. In 1952 the Pilkington Brothers in England developed the float glass process, in which a continuous 3.4-metre- (11-foot-) wide ribbon of glass floated over molten tin and both sides were fire finished, avoiding all polishing and grinding; this became the standard method of production. Pilkington also pioneered the development of structural glass mullions in the 1960s. In the 1950s the rise of air conditioning led to the marketing of tinted glass that would absorb and reduce solar gain, and in the 1960s reflective glass with thin metallic coatings applied by the vacuum plating process was introduced, also to reduce solar gain. Heat-mirror glass, which has a transparent coating that admits the short-wavelength radiation from the Sun but tends to reflect the longer-wavelength radiation from within occupied spaces, was introduced in 1984; when combined with double glazing its insulating value approaches that of a wall.

Curtain-wall construction

High-rise construction since 1945. *Use of steel and other metals.* The second great age of high-rise buildings began after the end of World War II, when the world economy and population again expanded. It was an optimistic time with declining energy costs, and architects embraced the concept of the tall building as a glass prism. This idea had been put forward by the architects Le Corbusier and Ludwig Mies van der Rohe in their visionary projects of the 1920s. These designs employed the glass curtain wall, a non-load-bearing "skin" attached to the exterior structural components of the building. The earliest all-glass curtain wall, which was only on a single street facade, was that of the Hallidie Building (1918) in San Francisco. The first multistory structure with a full glass curtain wall was the A.O. Smith Research Building (1928) in Milwaukee by Holabird and Root; in it the glass was held by aluminum frames, an early use of this metal in buildings. But these were rare examples, and it was not until the development of air conditioning, fluorescent lighting, and synthetic rubber sealants after 1945 that the glass prism could be realized.

The paradigm of the glass tower was defined by the United Nations Secretariat Building (1949) in New York City; Wallace Harrison was the executive architect, but Le Corbusier also played a major role in the design. The UN building, which featured a Weathermaster air-conditioning system and green-tinted glass walls, helped set the standard for tall buildings around the world. Several other influential buildings—such as Mies van der Rohe's 26-story 860–880 Lake Shore Drive Apartments (1951) in Chicago and Skidmore Owings & Merrill's 21-story Lever House (1952) in New York City—helped to further establish the technology of curtain walls. Perhaps the most important element was the development of extruded-aluminum mullion and muntin shapes to support the glass. Aluminum began to be produced in quantity in the United States by the Hall process in 1886; this process for separating the metal from the ore required large amounts of electricity, and declining energy costs after World War II influenced the development of this building technology. Aluminum forms a coating of transparent oxide that protects it against corrosion; this oxide layer can be artificially thickened and coloured through a process called anodizing. Anodized aluminum was first used in the windows of the Cambridge University Library in England in 1934. Aluminum became the principal material of curtain-wall framing because of its corrosion resistance and ease of forming by means of the extrusion process, in which the metal is forced through a series of dies to create complex cross-sectional shapes. Formed sheet aluminum is also used for opaque curtain-wall panels. Other metals used in curtain walls are stainless steel (a compound of 82 percent iron and 18 percent chromium) and so-called weathering steel, copper-bearing steel alloys that form an adherent oxide layer. The bronze curtain wall of Mies van der Rohe's Seagram Building (1954–58) in New York City proved to be an isolated example. Probably of equal importance in curtain-wall construction was the development of cold-setting rubbers during World War II; these form the elastic sealants that successfully seal the joints between glass and metal and between metal and metal against wind and rain. In the late 1970s the development of artificial diamonds made possible cutting tools that slice stone wafer-thin, and it became an important component of curtain walls.

The bundled tube

Following the development of the curtain wall, new forms of structure appeared in high-rise buildings. As environmental control systems increased in cost, economic pressures worked to produce more efficient structures. In 1961 the 60-story Chase Manhattan Bank Building, designed by Skidmore, Owings & Merrill, had a standard steel frame with rigid portal wind bracing, which required 275 kilograms of steel per square metre (55 pounds of steel per square foot), nearly the same as the Empire State Building of 30 years earlier. Economy of structure in tall buildings was demonstrated by the same firm only nine years later in the John Hancock Building in Chicago. It used a system of exterior diagonal bracing to form a rigid tube devised by the engineer Fazlur Khan; although the Hancock building is 100 stories and 343 metres (1,127 feet) high, its structure is so efficient that it required only 145 kilograms of steel per square metre (29 pounds per square foot). The framed tube, which Khan developed for concrete structures, was applied to other tall steel buildings. Khan used a steel system of nine bundled tubes of different heights—each 22.5 metres (75 feet) square with columns spaced at 4.5 metres (15 feet)—to form the structure of the 110-story Sears Tower (1974), also in Chicago (Figure 9). Its height of 443 metres (1,454 feet) has not been exceeded, and, although present technology would undoubtedly permit higher structures to be built, the large floor areas they provide seem to be unfavourable for economic considerations.

Use of reinforced concrete. Parallel to the development of tall steel structures, substantial advancements in high-rise structural systems of reinforced concrete have been made since 1945. The first of these was the introduction of the shear wall as a means of stiffening concrete frames against lateral deflection, such as results from wind or earthquake loads; the shear wall acts as a narrow deep cantilever beam to resist lateral forces. In 1958 the ar-

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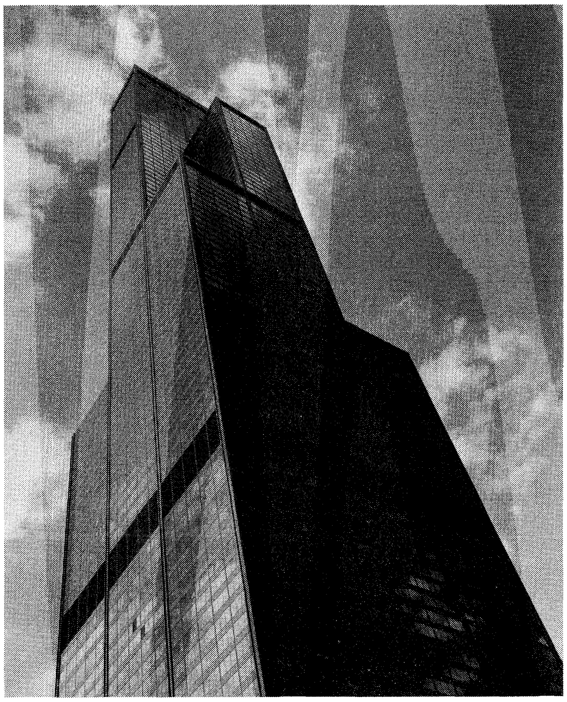


Figure 9: Bundled tube construction with termination of tubes at various heights, Sears Tower, Chicago, 1974.

chitect Milton Schwartz and engineer Henry Miller used shear walls to build the 39-story Executive House in Chicago to a height of 111 metres (371 feet). Of equal importance was the introduction of the perimeter-framed tube form in concrete by Fazlur Khan in the DeWitt–Chestnut Apartments (1963) in Chicago; the building rises 43 stories (116 metres, or 387 feet). Lateral stability was achieved by closely spaced columns placed around the building perimeter and connected together by deep beams. The next step in concrete high-rise construction was the combination of the perimeter-framed tube with a largely solid-walled interior tube or shear walls to give further lateral stability. This was employed by Eero Saarinen and Kevin Roche in the 35-story CBS Building (1964) in New York City, and the system was further developed by Khan in the 221-metre (725-foot) Shell Oil Building (1967) in Houston. Another new structural form in concrete was introduced by Khan in the 174-metre (570-foot) 780 Third Avenue Office Building (1983) in New York City. This is a framed tube with diagonal bracing achieved by filling in diagonal rows of window openings to create exterior bracing members; this is a very efficient system and may lead to yet taller buildings of this type. Three further innovations helped the rapid rise in height of concrete buildings. One was the development of lightweight concrete, using blast-furnace slag in place of stone as aggregate for floor construction; this reduced the density of the concrete by 25 percent, with a corresponding reduction in the loads the building columns needed to carry. The second was the increase in the ultimate strength of concrete used for columns. Third, the use of pumps to move liquid concrete to the upper floors of tall buildings substantially reduced the cost of placement.

Another important technique developed for concrete high-rise construction is slipforming. In this process, a continuous vertical element of planar or tubular form is continuously cast using a short section of formwork that is moved upward with the pouring process. Slipforming has been used to build a number of very tall structures in Canada, including several industrial chimneys 366 metres (1,200 feet) high and the world's tallest occupied structure, the Canadian National Railroad Tower in Toronto, which contains only an observation deck and a television antenna but has a total height of 553 metres (1,814 feet). Concrete has shown itself to be a serious competitor with steel in high-rise structures; it is now used for the great majority of tall residential buildings and for a substantial number of tall office buildings.

Postwar developments in long-span construction. After 1945 the dome and the shell vault continued to be the major forms of long-span structures. One innovation was the geodesic dome, which was devised by the architect and engineer R. Buckminster Fuller in the 1940s; in this form the ribs are placed in a triangular or hexagonal pattern and lie on the geodesic lines, or great circles, of a sphere. A very shallow spherical form with aluminum trussed members was used by Freeman Fox & Partners for the Dome Discovery built in London in 1951. Fuller's own patented forms were used in 1958 to build two large hemispheric domes 115.3 metres (384 feet) in diameter using steel tube members. These are used as workshops for the Union Tank Car Company in Wood River, Ill., and Baton Rouge, La. The largest geodesic dome is the Poliedro de Caracas, in Venezuela, built of aluminum tubes spanning 143 metres (469 feet).

The geodesic dome

Another form of steel trussed dome is the lamella dome, which is made of intersecting arches hinged together at their midpoints to form an interlocking network in a diamond pattern. It was used for the first two examples of the great covered sports stadiums built in the United States since the 1960s: the Harris County Stadium, or Astrodome, built in Houston, Texas, in 1962–64 with a span of 196 metres (642 feet) and the 207-metre- (678-foot-) diameter Superdome in New Orleans, La., designed by Sverdrup and Parcel and completed in 1973. The steel truss continued to be used and was extended to three dimensions to form space trusses. The longest span of this type was the Narita Hangar at Tokyo International Airport, which used a tied portal truss to span 190 metres (623 feet) supporting a space-truss roof spanning 90 metres (295 feet).

The concrete dome or shell developed rapidly in the 1950s. The St. Louis Lambert Airport Terminal (1954), designed by Hellmuth, Yamasaki and Leinweber, has a large hall 36.6 metres (120 feet) square, spanned by four intersecting thin-shell concrete barrel vaults supported at the four corners; the thickness of the shell varies from 20 centimetres (eight inches) at the supports to 11.3 centimetres (4.5 inches) at the centre. Another example is the King Dome, in Seattle, Wash., which covers a sports stadium with a thin single shell concrete parabolic dome stiffened with ribs 201 metres (661 feet) in diameter.

New forms of the long-span roof appeared in the 1950s based on the steel cables that had long been used in suspension bridges. One example was the U.S. Pavilion at the 1958 Brussels World's Fair, designed by the architect Edward Durell Stone. It was based on the familiar principle of the bicycle wheel; its roof had a diameter of 100 metres (330 feet), with a steel tension ring at the perimeter from which two layers of radial cables were tightly stretched to a small tension ring in the middle—the double layer of cables gave the roof stability against vertical movement. The Oakland–Alameda County Coliseum (1967), by Skidmore, Owings & Merrill, extended this system to 126 metres (420 feet) in diameter, but only a single layer of cables, stiffened by encasing ribs of concrete, connects the inner and outer rings.

Another system derived from bridge construction is the cable-stayed roof. An early example is the TWA Hangar (1956) at Kansas City, Mo., which shelters large aircraft under a double cantilever roof made of semicylindrical shells that reach out 48 meters (160 feet); deflection is reduced and the shells kept in compression by cables that run down from central shear walls to beams in the valleys between the shells. Another example of the cable-stayed roof is the McCormick Place West Exhibition Hall (1987) in Chicago, by Skidmore, Owings & Merrill. Two rows of large concrete masts rise above the roof, supporting steel trusses that span 72 metres (240 feet) between the masts and cantilever 36 metres (120 feet) to either side; the trusses are also supported by sets of parallel diagonal cables that run back to the masts.

A third form of long-span roof structures in tension are air-supported plastic membranes, which were devised by Walter Bird of Cornell University in the late 1940s and were soon in use for swimming pools, temporary warehouses, and exhibition buildings. The Osaka World's Fair of 1970 included many air-supported structures, the largest of which was the U.S. Pavilion designed by the engineers Geiger Berger Associates; it had an oval plan 138 × 79 metres (460 × 262 feet), and the inflated domed roof of vinyl-coated fabric was restrained by a diagonally intersecting network of steel cables attached to a concrete compression ring at the perimeter. The system of the Osaka Pavilion was adapted for two large sports stadiums built in the 1980s: the Silverdome at Pontiac, Mich., and the Hubert H. Humphrey Metrodome in Minneapolis. Air-supported structures are perhaps the most cost-effective type of structure for very long spans.

Air-supported structures

Building construction has settled into a period of relative calm after the explosive innovations of the 19th century. Steel, concrete, and timber have become fairly mature technologies, but there are other materials—such as fibre composites—that may yet play a major role in building.

Modern building practices

THE ECONOMIC CONTEXT OF BUILDING CONSTRUCTION

Buildings, like all economic products, command a range of unit prices based on their cost of production and their value to the consumer. In aggregate, the total annual value of building construction in the various national economies is substantial. In 1987 in the United States, for example, it was about 10 percent of the gross domestic product, a proportion that is roughly applicable for the world economy as a whole. In spite of these large aggregate values, the unit cost of buildings is quite low when compared to other products. In the United States in 1987, new building cost ranged from about $0.50 to $2.50 per pound. The lowest costs are for simple pre-engineered metal buildings, and the highest represent functionally complex buildings with many mechanical and electrical services, such as hospitals and laboratories. These unit costs are at the low end of the scale of manufactures, ranking with inexpensive foodstuffs, and are lower than those of most other familiar consumer products. This scale of cost is a rough index of the value or utility of the commodity to society. Food, although essential, is relatively easy to produce; aircraft, at the high end of the scale, perform a desirable function but do so with complex and expensive mechanisms that command much higher unit prices which reflect not only the materials and labour required to produce them but also substantial capital and research investments. Buildings fall nearer to food in value; they are ubiquitous and essential, yet the services consumers expect them to provide can be supplied with relatively unsophisticated technology and inexpensive materials. Thus there has been a tendency for building construction to remain in the realm of low technology, for there has been relatively little incentive to invest in research given consumer expectations.

Building and zoning codes

Within this general economic context, there are a number of specific parameters that affect the cost of buildings. First are government building codes, which are enacted to protect public health and safety; these take the form of both prescriptive and performance requirements. Structural requirements include description of the loads buildings must support, beginning with the constant everyday loads of building contents imposed by gravity and extending to the less frequent but more extreme loadings of wind and earthquake forces. These are specified on a statistical basis, usually the maximum expected to occur with a 100-year frequency. Safety factors for materials are specified to allow for accidental overloading and lapses of quality control. Economic considerations are also reflected; for example, buildings must perform well under normal gravity loads, but no code requires a building to resist direct exposure to the wind and low pressure effects of a tornado, for its cost would be prohibitive.

Planning and zoning requirements provide for height and floor area limitations and building setbacks from lot lines to ensure adequate light and air to adjoining properties. Zoning regulations also establish requirements for permitted building usages, parking spaces, and landscaping and even set standards for the visual appearance of buildings. Another example is requirements for building atmosphere conditions; these include minimum (but not maximum) temperatures and rates of air change to dilute odours and provide an adequate oxygen supply. Life-safety requirements include adequate stairways for emergency exits, emergency lighting, smoke detection and control systems, and fire-resistant building materials. Sanitation requirements include adequate numbers of plumbing fixtures and proper pipe sizes. Electrical requirements include wire sizes, construction requirements for safety, and location of outlets.

Beyond the government standards there are market standards, which reflect user expectations for buildings. One example is elevator systems; elevators are not required by building codes, but in the United States, for example, the number of elevators in office buildings is calculated based on a maximum waiting period of 30 seconds. Cooling of building atmospheres is also not required by code but is provided in climates and building types where the marketplace has shown it to be cost-effective.

Building systems and components are perceived as having two dimensions of value. One is the purely functional dimension: the structure is expected to resist loads, the roof must keep out rain. The other is the aesthetic or psychic dimension: stone is perceived as more durable than wood; an elevator system with a waiting time of 30 seconds is preferable to one with a waiting time of two minutes. For these perceived differences many users are willing to pay more. When symbolic buildings such as temples, cathedrals, and palaces play an important role in society, the aesthetic dimension is important in valuing buildings; for example, the Parthenon of Athens or Chartres Cathedral commanded a level of investment in their economies that might be roughly compared to the U.S. Apollo space program. But in most buildings, the functional dimension of value is dominant.

Because of its relatively low level of technology, wide geographic distribution, highly variable demand, and wide variety of building products, the building industry in industrialized countries is subdivided into many small enterprises. This lack of centralization tends to discourage research and keeps building components sturdy and simple, following well-tried formulas. Within this diversity there are a number of fairly well-defined markets based on building types; these include low-rise residential buildings, low-rise commercial, institutional, and industrial buildings, high-rise buildings, and long-span buildings.

A somewhat similar pattern is found in the Soviet bloc, although the building industry there is more centralized. There is also a much smaller low-rise residential market, with most new housing being provided in high-rise buildings.

In developing countries the major market is for low-rise residential buildings to house rapidly growing populations. Much of the construction is undertaken by local craftsmen using simple building products. Local timber is widely used, and masonry materials still include the ancient mud brick. More sophisticated long-span and high-rise technologies are found only in major cities.

BUILDING DESIGN AND CONSTRUCTION

Design programming. The design of a building begins with its future user or owner, who has in mind a perceived need for the structure, as well as a specific site and a general idea of its projected cost. The user, or client, brings these facts to a team of design professionals composed of architects and engineers, who can develop from them a set of construction documents that define the proposed building exactly and from which it can be constructed.

Building design professionals include those licensed by the state—such as architects and structural, mechanical, and electrical engineers—who must formally certify that the building they design will conform to all governmental codes and regulations. Architects are the primary design professionals; they orchestrate and direct the work of engineers, as well as many other consultants in such specialized areas as lighting, acoustics, and vertical transportation.

Building science

The design professionals draw upon a number of sources in preparing their design. The most fundamental of these is building science, which has been gradually built up over the past 300 years. This includes the parts of physical theory that relate to building, such as the elastic theory of structures and theories of light, electricity, and fluid flow. There is also a large compendium of information on the specific properties of building materials that can be applied in mathematical models to reliably project building performance. There is also a large body of data on criteria for human comfort in such matters as thermal environment, lighting levels, and sound levels that influence building design.

In addition to general knowledge of building science, the design team collects specific data related to the proposed building site. These include topographic and boundary surveys, investigations of subsoil conditions for foundation and water-exclusion design, and climate data and other local elements.

Concurrently with the collection of the site data, the design team works with the client to better define the often vague notions of building function into more precise and concrete terms. These definitions are summarized in a building space program, which gives a detailed written description of each required space in terms of floor area, equipment, and functional performance criteria. This document forms an agreement between the client and the design team as to expected building size and performance.

Design development. The process by which building science, site data, and the building space program are used by the design team is the art of building design. It is a complex process involving the selection of standard building systems, and their adaptation and integration, to produce a building that meets the client's needs within the limitations of government regulations and market standards. These systems have become divided into a number of clear sectors by the building type for which they are intended. The design process involves the selection of systems for foundations, structure, atmosphere, enclosure, space division, electrical distribution, water supply and drainage, and other building functions. These systems are made from a limited range of manufactured components but permit a wide range of variation in the final product. Once the systems and components have been selected, the design team prepares a set of contract documents, consisting of a written text and conventionalized drawings, to describe completely the desired building configuration in terms of the specified building systems and their expected performance. When the contract documents have been completed, the final costs of the building can usually be accurately estimated and the construction process can begin.

The general contractor

Construction. Construction of a building is usually executed by a specialized construction team; it is normally separate from the design team, although some large organizations may combine both functions. The construction team is headed by a coordinating organization, often called a general contractor, which takes the primary responsibility for executing the building and signs a contract to do so with the building user. The cost of the contract is usually an agreed lump sum, although cost-plus-fee contracts are sometimes used on large projects for which construction begins before the contract documents are complete and the building scope is not fully defined. The general contractor may do some of the actual work on the building in addition to its coordinating role; the remainder of the work is done by a group of specialty subcontractors who are under contract to the general contractor. Each subcontractor provides and installs one or more of the building systems—*e.g.,* the structural or electrical system. The subcontractors in turn buy the system components from the manufacturers. During the construction process the design team continues to act as the owner's representative, making sure that the executed building conforms to the contract documents and that the systems and components meet the specified standards of quality and performance.

LOW-RISE RESIDENTIAL BUILDINGS

Low-rise residential buildings include the smallest buildings produced in large quantities. Single-family detached houses, for example, are in the walk-up range of one to three stories and typically meet their users' needs with about 90 to 180 square metres (about 1,000 to 2,000 square feet) of enclosed floor space. Other examples include the urban row house and walk-up apartment buildings. Typically these forms have relatively low unit costs because of the limited purchasing power of their owners. The demand for this type of housing has a wide geographic distribution, and therefore most are built by small local contractors using relatively few large machines (mostly for earth moving) and large amounts of manual labour at the building site. The demand for these buildings can have large local variations from year to year, and small builders can absorb these economic swings better than large organizations. The building systems developed for this market reflect its emphasis on manual labour and its low unit costs. A proportion of single-family detached houses are "factory-built"; that is, large pieces of the building are prefabricated and then transported to the site, where considerable additional work is required to complete the finished product.

Foundations. All foundations must transmit the building loads to a stable stratum of earth. There are two criteria for stability: first, the soil under the foundations should be able to receive the imposed load without more than about 2.5 centimetres (one inch) of settlement and, second, the settlement should be uniform under the entire building. It is also important that the bottom of the foundation be below the maximum winter frost level. Wet soil expands as it freezes, and repeated freeze–thaw cycles can move the building up and down, leading to possible displacement and damage. Maximum frost depth varies with climate and topography. It can be as deep as 1.5 metres (five feet) in cold continental climates and is zero in tropical and some subtropical areas. The foundation systems for low-rise residential buildings are suitable for their light loads; nearly all are supported on spread footings, which are of two types—continuous footings that support walls and isolated pad footings that support concentrated loads. The footings themselves are usually made of concrete poured directly on undisturbed soil to a minimum depth of about 30 centimetres (12 inches). If typical continuous concrete footings are used, they usually support a foundation wall that acts either as a retaining wall to form a basement or as a frost wall with earth on both sides. Foundation walls can be built of reinforced concrete or masonry, particularly concrete block. Concrete blocks are of a standard size larger than bricks and are hollow, forming a grid of vertical planes. They are the least expensive form of masonry—using cheap but strong material—and their large size economizes on the labour required to lay them. Their appearance and weathering properties are inferior to those of fired masonry, but they are satisfactory for foundation walls. In some places timber foundation walls and spread footings are used. Excavation for foundations is the most highly mechanized operation in this building type; it is done almost entirely with bulldozers and backhoes.

Spread footings

Structural systems. *Timber frames.* In these small buildings the ancient materials of timber and masonry are still predominant in the structural systems. In North America, which has abundant softwood forests, light timber frames descended from the 19th-century balloon frame are widely used. These present-day "platform" frames are made of standard-dimension timbers, usually two or four centimetres (0.75 or 1.5 inch) thick, which are joined together by machine-made nails and other metal fasteners using hand tools (Figure 10).

The first step is to construct a floor, which rests on the foundation wall. A heavy timber sill is attached to the wall with anchor bolts, and on top of it are nailed the floor joists, typically 4 × 28 centimetres (1.5 × 11.25 inches) and spaced 40 centimetres (16 inches) apart. The span of the floor joists is usually about 3.6 metres (12 feet), which is the common maximum length of available timbers. The floor may need intermediate supports in the form of interior foundation walls or, if there is a basement, intermediate beams of wood or steel supported by the foundation walls and columns. For longer spans, floor trusses can be made, with members joined by nail grids or nailed plywood gussets or with wood chords and diagonal metal web members. On top of the joists is nailed plywood subflooring, which forms the deck and gives lateral stability to the floor plane.

The exterior bearing walls are made of 4 × 9-centimetre (1.5 × 3.5-inch; "2 × 4") timber verticals, or studs, spaced 40 or 60 centimetres (16 or 24 inches) apart, which rest on a horizontal timber, or plate, nailed to the floor platform and support a double plate at the top. The walls are sheathed on the outside with panels of plywood or particleboard to provide a surface to attach the exterior cladding and for lateral stability against wind. Plywood and particleboard are fabricated in panels of standard sizes. Plywood is made of thin layers of wood, rotary-cut from logs and glued together with the wood grain running perpendicularly in adjoining layers. Particleboard consists of fine wood chips mixed together in an adhesive matrix and allowed to harden under pressure. On top of the wall plate is placed either a second floor or the roof.

Since most of the roofing materials used in these buildings are not fully watertight, the roofs must have sloped surfaces to rapidly drain off rainwater. Sloped forms are created by two methods. The traditional method uses joists similar to those of floor construction to span between exterior walls. Rafters are nailed to the ends of each joist and the rafters meet at a central ridge member, forming a triangular attic space. Where no attic space is needed, it has proved more economical to span the roof with triangular trusses with interior web members. These roof trusses are usually made of narrow timbers joined by nails, glue, or metal connectors, and they are often prefabricated in a workshop. Plywood or particleboard sheathing is then nailed to the roof surfaces to receive the roofing

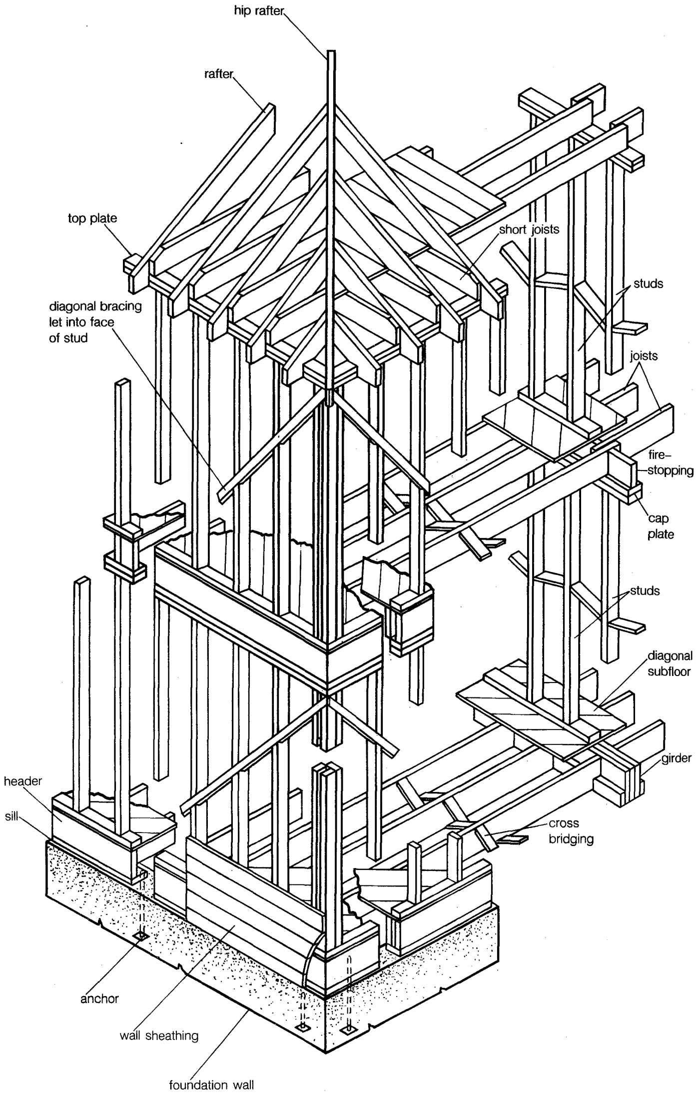


Figure 10: Platform framing for light timber construction, axonometric view.

Adapted from C.G. Ramsay and H.R. Sleeper, *Architectural Graphic Standards* (1970), John Wiley and Sons

and to provide lateral stability, making the entire frame into a rigid box.

Problems of timber framing

Light timber frames are quite flammable, but small one- or two-story buildings are easy to evacuate in case of a fire, and building codes permit the use of these frames with such features as fire-resistant gypsum board on the interiors and fire-stops (short wooden members) between the studs. Timber structures are attacked by certain species of insects—such as termites and carpenter ants—as well as certain fungi, particularly in warm, moist climates. Wood can be chemically treated to discourage these attacks; other precautions include raising the timber above the ground and keeping it dry.

Masonry walls. Structural masonry walls are also used in this building type, primarily in multistory buildings, where they offer greater load-bearing capacity and fire resistance. Brick and concrete block are the major materials, brick being favoured for exterior surfaces because of its appearance and durability. Solid brick walls are rarely used, due to the higher labour and material costs; composite walls of brick and block or block alone are common. Cavity walls are used in colder climates; in these, two wythes (vertical layers) of masonry are built on either side of a layer of rigid insulation. The wythes are joined together by steel reinforcement that runs through the insulation and is laid in the horizontal masonry joints at intervals. Cavity walls have a heat-flow rate that is 50 percent of that of a solid wall. Timber floor and roof construction, similar to balloon framing, is used with masonry construction; and there is also some use of precast prestressed hollow concrete panels, which are fireproof and can span up to nine metres (30 feet).

Enclosure systems. Enclosure systems for this building type are varied. For roofs, traditional wood shingles or, more commonly, felt asphalt shingles are used, as are semicylindrical clay tiles and standing-seam metal roofs. Rainwater from roofs is usually caught in metal gutters and directed to exterior downspouts that discharge onto splash blocks or into underground drains connected to storm sewers.

The wall surfaces of low-rise residential buildings are clad with a range of different materials. Traditional wood elements such as shingles and horizontal shiplap, or clapboard siding, are used on light timber frames as are vertical tongue-and-groove siding and boards and battens. Aluminum and vinyl sidings have been adapted from these wooden forms. Brick and stone veneer are also applied over timber and anchored to it with metal fasteners. Cement plaster, or stucco, is another traditional material used to enclose both timber and masonry structures, and its semiliquid application allows great plasticity of form. A more recent development is a very thin synthetic resin stucco applied directly to the surface of rigid plastic foam insulation.

Insulation

Insulation, which slows the rate of heat transfer through the enclosure, is usually applied at all exterior building surfaces that are exposed to air. There are two major types of insulation, rigid and nonrigid. Rigid insulations are primarily plastic foams (the dead air in the foam cells is the true insulator), which vary in thickness from 2.5 to five centimetres (one to two inches). They include styrofoam, used primarily below grade behind frost walls due to its low fire resistance; urethane foam; isocyanurate foam, which has the best fire resistance; and foam glass. Nonrigid insulations are usually made of fibre—glass fibre being the most common—often with a foil-backed paper on one side. Fibre insulations are made in thicknesses up to 23 centimetres (9.25 inches). The effectiveness of an insulation material is measured in terms of its heat-transfer rate, or U-value, often expressed as the number of BTUs passing through a given unit of insulating material each hour at an expressed temperature differential across the material. Low U-values indicate good insulating properties of the material. U-value is an inverse function of thickness, so that there is a limit to the cost-effectiveness of increasing the amount of insulation on a surface. Rigid insulation panels are applied to vertical wall sheathing and the surfacing material is fastened through the insulation, or it is applied to horizontal roof decks. Glass fibre is usually applied in the spaces between wall studs and between roof joists or the bottom chords of roof trusses.

Most low-rise residential buildings have a limited number of transparent openings in their exteriors, because of the traditional requirements of interior privacy and the relatively higher cost of windows compared to opaque walls. The traditional wooden frames of domestic windows are often clad in extruded vinyl or aluminum cladding, and frames made entirely of extruded aluminum are common. Residential windows are a major means of ventilation, and there are a variety of operating actions for their movable sections: sliding or double-hung windows are still the major form, but hinged types—including casement, hopper, and awning forms—are also used. Sliding glass panel doors are also used, particularly in warmer regions. Glazing is still largely of clear glass. Double glazing, with two panes bonded to a metal tubular separator that contains a desiccant, is cost-effective in northern climates, but triple glazing is used commonly only in regions above about 55° to 60° latitude. A recent development is heat-mirror glass, in which a low-emissivity coating enhances the relative opacity of the glass to infrared radiation and slows the rate of internal heat loss in winter.

Interior finishes. Interior finishes and space-division systems define the living spaces within residential buildings with a range of both natural and synthetic materials. The most widely used wall finish is gypsum board, a prefabricated form of traditional wet plaster. Wet gypsum plaster is cast between paper facings to form large panels that are nailed to light timber or metal frameworks. The joints between the panels are filled with a hard-setting resin compound, giving a smooth seamless surface that has considerable fire resistance. Gypsum board forms the substrate to which a number of other materials, including thin wood-veneered plywood and vinyl fabrics, can be applied with adhesives. In wet areas such as kitchens and bathrooms, water-resistant gypsum board is used, sometimes with the addition of adhesive-applied ceramic tile.

Doors in residential buildings are usually of the hollow-core type, with thin veneers of wood glued over a honeycomb paper core and solid wood edge strips; door frames are typically made of machined timber shapes. Plastic laminates bonded to particleboard are extensively used for built-in cabinets and countertops. The most common floor finish is carpeting, most of which is now made of synthetic fibres, displacing the traditional wool and cotton. It can be easily maintained, and its soft visual and tactile texture, as well as its sound-absorbing qualities, make it attractive for residential use. Hardwoods—primarily oak, birch, and maple—are also used for floors, both in the traditional narrow planks nailed to plywood decks and as prefabricated parquet elements, which are applied with adhesives. In wet or hard-use areas vinyl-composition tiles or ceramic tiles are used.

Mechanical and electrical systems. *Plumbing.* Domestic water-supply systems for low-rise residential buildings have two sources, either municipal water-distribution systems or, where these are not available, wells that are drilled to underground aquifers which are free of contamination. Water is drawn from the wells with small submersible electric pumps, which are lowered through the well casing to the intake. Underground exterior water-supply pipes are usually cast-iron with threaded connections to contain the pressures applied to the fluid, which is typically sufficient to raise it four stories. Within the building, copper tubing with soldered connections is used for distribution because of its corrosion resistance and ease of fabrication; in some areas plastic pipe is also used. The domestic water supply is divided into cold and hot systems, the cold water being piped directly to the fixtures. The hot-water system first draws the supply through a hot-water heating tank, which raises its temperature to about 60° C (140° F) using electric resistance or gas heat. Domestic water heaters that use solar radiation to heat water in coils exposed to the sun on a glass-covered black metal plate (flat-plate solar collectors) are found in areas where there is ample sunshine and relatively high energy costs. The hot water is then distributed from the heater to the fixtures in a recirculating loop pipe system, in which gravity and temperature differentials maintain a constant temperature in period of low demand.

Domestic plumbing service

The primary residential use of water is in the bathroom, which typically includes a bathtub of cast iron or pressed steel with a ceramic porcelain coating (although fibreglass-reinforced resin is also used), a ceramic lavatory, and a ceramic tank-type water closet. The bath and lavatory are supplied with hot and cold water through faucets with lever or screw-type valve controls. The valve of the water closet supply is also lever-operated and relies on the gravity power of the water in the tank for its flushing action. Shower baths are also common, often incorporated into bathtub recesses or in a separate compartment finished with ceramic tile. In some countries a bidet is included.

Other widely used plumbing fixtures include kitchen sinks, usually of cast iron or pressed steel with a ceramic porcelain coating, or of stainless steel; automatic dishwashing machines; and automatic washing machines for laundry. Kitchen sinks can be fitted with garbage disposals, which grind solid waste into a fluid slurry that is flushed out with wastewater. Where the possibility of back siphonage of wastewater into the water supply exists, a vacuum breaker must be provided at the supply to prevent this happening, but most domestic plumbing fixtures are designed to avoid this possibility.

Drainage systems to remove wastewater are made of cast-iron pipe with threaded joints or bell-and-spigot joints sealed with molten lead or with plastic pipe with solvent-welded joints. The waste pipe of every plumbing fixture is provided with a semicircular reverse curve, or trap, which remains constantly filled with water and prevents odours from the drainage system from escaping into occupied spaces. Immediately downstream from each trap is an opening to a vent pipe system, which lets air into the drainage system and protects the water seals in the traps from removal by siphonage or back pressure. When wastewater leaves the building, it is drained through a backflow-prevention valve and into underground ceramic pipes. It then flows by gravity to either a private sewage treatment plant, such as septic tank and tile field, or to the public sewer system. If the discharge level of the wastewater is below the level of the sewer, a sewage ejector pump is required to raise the wastewater to a higher level, where gravity carries it away.

Heating and cooling systems. Atmosphere-control systems in low-rise residential buildings use natural gas, fuel oil, or electric resistance coils as central heat sources; usually the heat generated is distributed to the occupied spaces by a fluid medium, either air or water. Electric resistance coils are also used to heat living spaces directly with radiant energy. Forced-air distribution moves the heat-bearing air through a treelike system of galvanized sheet-metal ducts of round or rectangular cross section; electric-powered fans provide a pressure differential to push the air from the heat source (or furnace) to the living spaces, where it is expelled from grills located in the walls or floors. The negative pressure side of the fan is connected to another treelike system of return air ducts that extract air from living spaces through grills and bring it back to the furnace for reheating. Fresh outside air can be introduced into the system airstream from an exterior intake, and odour-laden interior air can be expelled through a vent, providing ventilation, usually at the rate of about one complete air change per hour. To conserve energy, air-to-air heat exchangers can be used in the exhaust–intake process. The heated air is usually supplied in constant volume, and the ambient temperature is varied in response to a thermostat located in one room. Central humidity control is rarely provided in this building type.

Forced-air heating

Another common heating system is the radiant hot-water type. The heat source is applied to a small boiler, in which water is heated and from which it is circulated by an electric pump in insulated copper pipes similar to a domestic hot-water system. The pipes can be connected to cast-iron or finned tube steel radiators within the living spaces. The radiators are placed near the areas of greatest heat loss (such as windows or outside walls) where their radiant energy heats the surrounding air and creates a convection cycle within the room, producing a roughly uniform temperature within it. The hot water can also be conducted through narrow pipes placed in a continuous looping pattern to create a large radiant surface; this pattern of pipes may be cast in a concrete floor slab or placed above a ceiling to heat the adjoining living space. Temperature control in hot-water systems uses a thermostat in the living space to adjust the pumped flow rate of the water to vary the heat supplied.

Radiant electric resistance heating systems use coils in baseboard units in the rooms, which create convection cycles similar to hot-water radiators, or resistance cables in continuous looped patterns embedded in plaster ceilings. Local temperature control can be much more precise with electric heating, because it is possible to install a thermostatically controlled rheostat to vary the energy output of relatively small sections of baseboard units or cable.

A type of space heating that is increasing in use in residential buildings is passive solar radiation. On sunny winter days, south-facing windows let in substantial amounts of energy, often enough to heat the entire building. Wood-burning fireplaces with chimneys are still widely provided in residential buildings, but their use is mostly for aesthetic effect.

The cooling of atmospheres in low-rise residential buildings is often done locally with unit air conditioners, which penetrate the exterior wall of the space to be cooled; this permits the intake of fresh air when desired and the ejection of heat pumped from the space to the exterior air. Less often, forced-air heating systems have cooling coils

Air-conditioning systems

introduced into the airstream to provide a centrally cooled interior. A compressive cooling process is used, similar to that in a domestic refrigerator. A refrigerant, which is a liquid at room temperature, is pumped through a closed system of coiled copper tubes. An electric pump maintains a low pressure in the cooling coils, and the liquid refrigerant passes through an expansion valve from a region of high pressure to the low-pressure coils. This change in pressure results in a phase change of the refrigerant; it turns from a liquid into a gas and absorbs heat in the process, just as water absorbs heat when it is boiled and converted into steam. The heat absorption of the liquid-to-gas transition cools air passing over the cooling coils. The cooled air is circulated through the building by the forced-air system. When the low-pressure gaseous refrigerant leaves the cooling coils, it goes through the pump and is pressurized. The refrigerant travels through condensing coils, which are located outside the building; there the phase change is reversed as the gas turns to a high-pressure liquid and liberates heat to the exterior air passing over the condensing coils. The liquid refrigerant returns to the expansion valve to repeat the cooling cycle. The refrigeration machine is thus a "heat pump" that moves heat out of the building to the exterior atmosphere. Heat pumps can also be run in reverse in the winter months to pump heat from the outside air into the building interior; they work best in mild climates with fairly warm winter temperatures. The use of heat pumps in cold climates poses many difficult technological problems.

Interior atmospheres are also ventilated by operating windows, as well as by unintended leakage at all types of exterior openings. Bathrooms, kitchens, and laundries generate odours and heat and often have separate exhaust systems powered by electric fans that are operated intermittently as required. Residential atmosphere quality is also protected by the smoke detector, which sounds an alarm to warn of possible danger when smoke reaches even a very low level in living spaces.

Electrical systems. Electrical systems in residential buildings are supplied from public utility power grids, starting from a step-down transformer near the building that reduces the high line voltage to a safer level. An underground or overhead cable from the transformer leads to the building, where it is connected to a meter that records the energy used by the subscriber. Immediately beyond the meter is a fused main switch to protect the building against an accidental power surge from the grid. The main service is then broken down into a number of circuits by a panelboard, each circuit having a fused switch. From the panelboard the wires of each circuit distribute the electricity to different areas of the building. The wires are usually copper, although aluminum is also used, and are covered with thermoplastic insulation. The wires must be contained in conduit, which is either metal or plastic tubing, to protect against damage and reduce the possibility of fire in the case of accidental overloading of the wires. Conduits are usually concealed in finished spaces within the framing of partition walls or above ceilings and terminate in junction boxes flush with a wall surface. The junction boxes contain terminal devices such as the convenience outlet, control switches, or the connection point for built-in light fixtures.

Residential lighting

Residential lighting is provided primarily by movable incandescent fixtures plugged into convenience outlets, but there is often built-in lighting in kitchens, bathrooms, corridors, and closets, mostly of the incandescent type. There is also some use of fluorescent lighting, particularly in built-in fixtures. Overall interior light levels in residential uses are low, about 20–40 footcandles. Exterior lighting is used for entrances, walkways, and exterior living spaces.

The power densities of dwelling units are fairly low and are declining because of the increased use of fluorescent lighting fixtures and improvement of efficiency in electrical appliances. The decline in power consumption enhances the prospect of the widespread appearance of dwellings—particularly detached houses—with their own independent electric power generation and storage systems, unconnected to public utility grids. Photovoltaic cells, which convert sunlight directly into electricity, in combination with storage batteries can offer these residences a new kind of energy autonomy.

LOW-RISE COMMERCIAL, INSTITUTIONAL, AND INDUSTRIAL BUILDINGS

The size of buildings in the commercial, institutional, and industrial market segment ranges from a few hundred to as much as 45,000 square metres (500,000 square feet). All of these buildings have public access and exit requirements, although their populations may differ considerably in density. The unit costs are generally higher than those for dwellings (although those of simple industrial buildings may be lower), and this type includes buildings with the highest unit cost, such as hospitals and laboratories. Residential buildings are fairly static in their function, changing only at long intervals. By contrast, most commercial, institutional, and industrial buildings must respond to fairly rapid changes in their functions, and a degree of flexibility is required in their component systems. In addition, these buildings are built by contractors who utilize heavy mechanized equipment not only for foundations (pile drivers and caisson augers) but also for lifting heavy components (a wide variety of cranes and hoists). Semimanual machines such as cement finishers, terrazzo grinders, and welding generators are also used, but a large percentage of the work is done manually; the human hand and back remain major instruments of the construction industry, well adapted to the nonrepetitive character of building.

Foundations. The foundations in these buildings support considerably heavier loads than those of residential buildings. Floor loadings range from 450 to 1,500 kilograms per square metre (100 to 300 pounds per square foot), and the full range of foundation types is used for them. Spread footings are used, as are pile foundations, which are of two types, bearing and friction. A bearing pile is a device to transmit the load of the building through a layer of soil too weak to take the load to a stronger layer of soil some distance underground; the pile acts as a column to carry the load down to the bearing stratum. Solid bearing piles were originally made of timber, which is rare today; more commonly they are made of precast concrete, and sometimes steel H-piles are used. The pile length may be a maximum of about 60 metres (200 feet) but is usually much less. The piles are put in place by driving them into the ground with large mechanical hammers. Hollow steel pipes are also driven, and the interiors are excavated and filled with concrete to form bearing piles; sometimes the pipe is withdrawn as the concrete is poured. An alternative to the bearing pile is the caisson. A round hole is dug to a bearing stratum with a drilling machine and temporarily supported by a steel cylindrical shell. The hole is then filled with concrete poured around a cage of reinforcing bars; and the steel shell may or may not be left in place, depending on the surrounding soil. The diameter of caissons varies from one to three metres (three to 10 feet). The friction pile of wood or concrete is driven into soft soil where there is no harder stratum for bearing beneath the site. The building load is supported by the surface friction between the pile and the soil.

Bearing piles and caissons

When the soil is so soft that even friction piles will not support the building load, the final option is the use of a floating foundation, making the building like a boat that obeys Archimedes' principle—it is buoyed up by the weight of the earth displaced in creating the foundation. Floating foundations consist of flat reinforced concrete slabs or mats or of reinforced concrete tubs with walls turned up around the edge of the mat to create a larger volume.

If these buildings do not have basements, in cold climates insulated concrete or masonry frost walls are placed under all exterior nonbearing walls to keep frost from under the floor slabs. Reinforced concrete foundation walls for basements must be carefully braced to resist lateral earth pressures. These walls may be built in excavations, poured into wooden forms. Sometimes a wall is created by driving interlocking steel sheet piling into the ground, excavating on the basement side, and pouring a concrete wall against it. Deeper foundation walls can also be built by the slurry wall method, in which a linear series of closely spaced caissonlike holes are successively drilled, filled with concrete, and allowed to harden; the spaces between are excavated by special clamshell buckets and also filled with concrete. During the excavation and drilling operations, the holes are filled with a high-density liquid slurry, which braces the excavation against collapse but still permits extraction of excavated material. Finally, the basement is dug adjoining the wall, and the wall is braced against earth pressure.

Structural systems. *Timber.* The structures of these buildings are mostly skeleton frames of various types, because of the larger spans their users require and the need for future flexibility. Timber is used, but on a much-reduced scale compared to residential buildings and primarily in regions where timber is readily available. The public nature of commercial and institutional buildings and the hazards of industrial buildings generally require that they be of noncombustible construction, and this largely excludes the use of light timber frames. Heavy timber construction can be used where the least dimensions of the members exceed 14 centimetres (5.5 inches); when timbers are this large they are charred but not consumed in a fire and are considered fire-resistant. Because most harvested trees are fairly small, it is difficult to obtain solid heavy timbers, and most large shapes are made up by glue laminating smaller pieces. The synthetic glues used are stronger than the wood, and members with cross sections up to 30 × 180 centimetres (12 × 72 inches) are made; these may be tapered or otherwise shaped along their length. Skeletons of glue-laminated beams and columns, joined by metal connectors, can span 30 to 35 metres (100 to 115 feet). Heavy decking made of tongue-and-groove planks up to 9.4 centimetres (3.75 inches) thick is used to span between beams to support floors and roofs.

Steel. Steel is a major structural material in these buildings (Figure 11). It is a strong and stiff material and yet relatively inexpensive, and it can be quickly fabricated and erected, which saves construction time. Although steel is noncombustible, it starts to lose strength when heated above 400° C (750° F), and building codes require it to be fireproofed in most multistory buildings; in small and low-hazard buildings, however, it can be left unprotected.

Types of structural steel

Nearly all structural steel—including sheets, round or square bars, tubes, angles, channels, and I beam or wide flange shapes—is formed by the hot-rolling process. Steel roof and floor deck panels are fabricated from sheet metal by further cold-rolling into corrugated profiles four to eight centimetres (1.5 to three inches) deep and 60 centimetres (24 inches) wide. They are usually welded to the supporting steel members and can span up to 4.5 metres (15 feet). The lightest and most efficient structural shape is the bar (or open web) joist, a standard truss made with angles for the top and bottom chords, joined by welding to a web made of a continuous bent rod. It is used almost exclusively to support roofs and can span up to 45 metres (150 feet). The standard rolled shapes are frequently used as beams and columns, the wide flange, or W shape, being the most common. The widely separated flanges give it the best profile for resisting the bending action of beams or the buckling action of columns. W shapes are made in various depths and can span up to 30 metres (100 feet). Where steel beams support concrete floor slabs poured onto a metal deck, they can be made to act compositely with the concrete, resulting in considerable economies in the beam sizes.

The connections of steel shapes are of two types: those made in the workshop and those made at the building site. Shop connections are usually welded, and site or field connections are usually made with bolts due to the greater labour costs and difficulties of quality control in field welding. Steel columns are joined to foundations with base plates welded to the columns and held by anchor bolts embedded in the concrete. The erection of steel frames at the building site can proceed very rapidly, because all the pieces can be handled by cranes and all the bolted connections made swiftly by workers with hand-held wrenches.

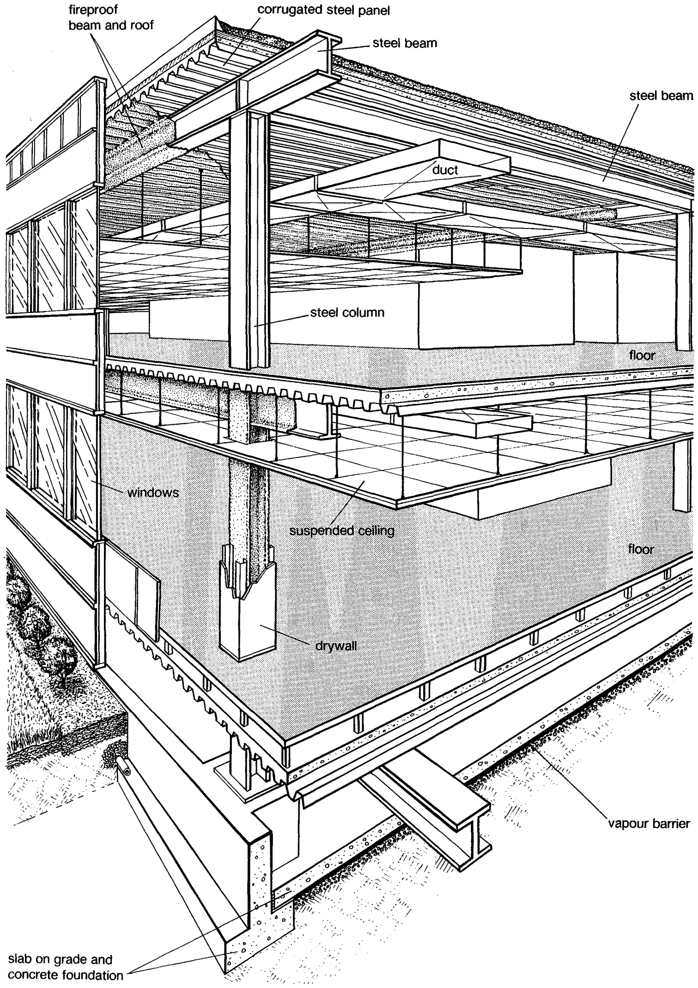


Figure 11: Low-rise steel-frame office building, corner perspective.

Adapted from R. Rush (ed.), *Building Systems Integration Handbook* (1986), John Wiley and Sons

A large proportion of steel structures are built as prefabricated, pre-engineered metal buildings, which are usually for one-story industrial and commercial uses. They are manufactured by companies that specialize in making such buildings of standard steel components—usually rigid steel bents or light trusses—which are assembled into frames and enclosed with corrugated metal siding. The configurations can be adapted to the needs of individual users. The metal building industry is a rare example of a successful application of prefabrication techniques in the construction industry in the United States, where its products are ubiquitous in the suburban and rural landscape.

Concrete. Reinforced concrete is also a major structural material in these buildings. Indeed, outside of North America and western Europe, it is the dominant industrialized building material. Its component parts are readily available throughout the world at fairly low cost. Portland cement is easily manufactured by burning shale and limestone; aggregates such as sand and crushed limestone can be easily obtained. Steel minimills, which use scrap iron to feed their electric furnaces, can mass-produce reinforcing bars for regional use. In industrialized countries the mixing and delivery of liquid concrete to building sites has been mechanized with the use of central plants and mixing trucks, and this has substantially reduced its cost. In barely 100 years, reinforced concrete has risen from an experimental material to the most widespread form of building construction.

In situ and precast concrete

There are two methods of fabricating reinforced concrete. The first is to pour the liquid material into forms at the building site; this is so-called in situ concrete. The other method is called precast concrete, in which building components are manufactured in a central plant and later brought to the building site for assembly. The components of concrete are portland cement, coarse aggregates such as crushed stone, fine aggregates such as sand, and water.

In the mix, water combines chemically with the cement to form a gel structure that bonds the stone aggregates together. In proportioning the mix, the aggregates are graded in size so the cement matrix that joins them together is minimized. The upper limit of concrete strength is set by that of the stone used in the aggregate. The bonding gel structure forms slowly, and the design strength is usually taken as that occurring 28 days after the initial setting of the mix. Thus there is a one-month lag between the time in situ concrete is poured and the time it can carry loads, which can significantly affect construction schedules.

In situ concrete is used for foundations and for structural skeleton frames. In low-rise buildings, where vertical gravity loads are the main concern, a number of framing systems are used to channel the flow of load through the floors to the columns for spans of six to 12 metres (20 to 40 feet). The oldest is the beam and girder system, whose form was derived from wood and steel construction: slabs rest on beams, beams rest on girders, and girders rest on columns in a regular pattern. This system needs much handmade timber formwork, and in economies where labour is expensive other systems are employed. One is the pan joist system, a standardized beam and girder system of constant depth formed with prefabricated sheet-metal forms. A two-way version of pan joists, called the waffle slab, uses prefabricated hollow sheet-metal domes to create a grid pattern of voids in a solid floor slab, saving material without reducing the slab's strength. The simplest and most economical floor system is the flat plate, where a plain floor slab about 20 centimetres (eight inches) thick rests on columns spaced up to 6.7 metres (22 feet) apart. If the span is larger, the increasing load requires a local thickening of the slab around the columns. When these systems are applied to spans larger than nine to 12 metres (30 to 40 feet), a technique called posttensioning is often used. The steel reinforcing takes the form of wire cables, which are contained in flexible tubes cast into the concrete. After the concrete has set and gained its full strength, the wires are permanently stretched taut using small hydraulic jacks and fastening devices, bending the entire floor into a slight upward arch. This reduces deflection, or sagging, and cracking of the concrete when the service load is applied and permits the use of somewhat shallower floor members. Concrete columns are usually of rectangular or circular profile and are cast in plywood or metal forms. The reinforcing steel never exceeds 8 percent of the cross-sectional area to guard against catastrophic brittle failure in case of accidental overloading.

Prestressing

Precast concrete structural members are fabricated under controlled conditions in a factory. Members that span floors and roofs are usually pretensioned, another prestressing technique, which is similar in principle to posttensioning. The reinforcement is again steel wire, but the wires are put into tension (stretched) on a fixed frame, formwork is erected around the taut wires, and concrete is poured into it. After the concrete has set and gained its full strength, the wires are cut loose from the frame. As in posttensioning, this gives the precast floor members a slight upward arch, which reduces deflection and permits the use of shallower members. Precast prestressed floor elements are made in a number of configurations. These include beams of rectangular cross section, hollow floor slabs 15 to 30 centimetres (six to 12 inches) deep and spanning up to 18 metres (60 feet), and single- and double-stem T shapes up to 1.8 metres (six feet) deep and spanning up to 45 metres (150 feet). Precast concrete columns are usually not prestressed and have projecting shelves to receive floor members. At the building site, precast members are joined together by a number of methods, including welding together metal connectors cast into them or pouring a layer of in situ concrete on top of floor members, bonding them together. Precast prestressed construction is widely used, and it is the dominant form of construction in the Soviet Union and eastern Europe.

Masonry. Masonry finds only a limited structural use in these buildings. Concrete block walls with brick facing and punched openings (discrete windows entirely surrounded by the facing material) spanned by concealed steel lintels can be used for exterior bearing walls where the interior is a skeleton frame of steel or timber. The use of interior bearing walls so greatly reduces the flexibility needed in these buildings that they are only rarely found.

Enclosure systems. Enclosure systems in these buildings range from rather simple forms in industrial uses to quite sophisticated assemblies in the commercial and institutional sectors. Most have in common the use of flat roofs with highly water-resistant coverings, the traditional one being a built-up membrane of at least four layers of coal-tar pitch and felt, often weighted down with a gravel ballast. Such roofs are pitched at slopes of 1 : 100 to 1 : 50 toward interior drains. In recent years the single-ply roof, made of plastic membranes of various chemistries, has found wide application. The seams between the pieces of membrane are heat- or solvent-welded together, and they are either ballasted with gravel or mechanically fastened to the underlying substrate, which is usually rigid foam insulation. Sometimes standing-seam sheet-metal roofs are also used; the best quality is continuously welded stainless steel.

Types of glazing

The choice of transparent surfaces in these enclosures is based on three major considerations: conductive heat transfer, radiant energy transfer, and safety. All the transparent materials used in the low-rise residential sector are found, plus a number of others. In buildings with fully controlled atmospheres, double glazing is common to reduce heat transfer and both interior and exterior condensation on the glass. Commercial and institutional buildings tend to have large internal sources of heat gain, such as people and lighting, so it is desirable to exclude at least some solar gain through the transparent surfaces to reduce energy consumption in cooling. This can be done by reducing the light transmission or shading coefficient of the glass by integrally tinting it in various colours; grey, bronze, and green are common tints. This can also be accomplished by vacuum-plating partial reflective coatings of varying densities to an inner surface of double glazing; this can reflect up to 90 percent of the incident energy. Two kinds of reflecting metal are used: aluminum, which is silver in tone, and rubidium, which is gold-toned. These coatings are perceived as strong tints when the outside world is viewed through them by day: grey for aluminum and green for rubidium.

Skylights or horizontal transparent surfaces have found wide application in these types of buildings. These installations range from purely functional daylighting in industrial uses to elaborate aesthetic forms in commercial structures. In horizontal applications, and in vertical walls where people might blunder into glazed panels, safety glazing is required. Safety glazing is of four types: certain plastics that are flexible and difficult to break; wire-embedded glass, which holds together when broken; tempered glass, which is very strong and breaks into tiny and relatively harmless fragments; and laminated glass, which consists of two layers of glass heat-welded together by an intermediate plastic film. Laminated glass can also be made with tinted lamination film, producing many colours not available in integrally coloured glass.

Because many of these buildings have skeleton structures, their vertical surfaces are enclosed in nonstructural curtain walls that resist wind forces and provide weatherproofing. Curtain walls are of several types; the most common is one supported by a metal (typically aluminum) gridwork attached to the building structure. The vertical members, called mullions, are attached to the building at every floor and are spaced 1.5 to three metres (five to 10 feet) apart; the horizontal members, called muntins, are attached between the mullions. The rectangles between the grid of mullions and muntins are filled with transparent or opaque panels. The transparent surfaces can be any of those just described, and the opaque panels include opaque coloured glass, painted or anodized aluminum sheets, porcelain enameled steel sheets, fibreglass-reinforced cement, and stone wafers of granite, marble, or limestone cut with diamond-edged tools. All of these materials are usually backed up by rigid insulation to slow heat transfer. Metal sandwich panels are also used for economy of material; two thin layers of metal are separated by a core of different material, often with a high U-value for insulating effect.

The separation of the thin layers of strong metal greatly increases the overall stiffness of the panel. The joints between panels and the supporting grid are weatherproofed with elastomeric sealants (cold-setting synthetic rubbers) or by prefabricated rubber gaskets. In glazed areas of curtain walls, mullions of structural glass are an alternative to metal mullions; they are more expensive, but they give an effect of greater transparency where this is desired.

Another type of curtain wall is the panel type. It has no gridwork of mullions and muntins but is made of large prefabricated rigid panels connected to the floors and spanning between them, with transparent openings made as holes cut out of the panel. The panels can be made of precast concrete, aluminum, or steel, often in sandwich form; elastomeric sealants are used to close the joints.

The finishes of metals in curtain walls include anodizing of aluminum, an electrolytic process that builds up the natural colourless oxide of aluminum into a thick adherent layer; it often includes the introduction of colour into the oxide layer itself. Durable paint coatings (with lifetimes of up to 40 years) can be applied to the metal in the factory; more conventional paints that must be renewed at shorter intervals are also used.

Walls and doors

Interior finishes. *Partitions.* Space-division systems in these buildings make use of gypsum board partitions, usually applied to a framework of formed sheet-metal members attached to the building structure. They are readily demolished and rebuilt at relatively low cost, meeting the need for flexibility in such buildings. They are often used for fire-resistive protective enclosures, for which a number of layers are laminated to achieve the specified fire resistance. Transparent and translucent partitions are also used, with different types of glass set in metal frames. Office buildings may contain prefabricated movable metal partitions, which typically use metal sandwich panel construction to create panels with both transparent and opaque surfaces as well as doors. These partitions are expensive compared with gypsum board and must be moved often to justify the greater initial cost. Concrete block is used in unfinished spaces and for fire-resistive partitions. Glazed ceramic block or ceramic tile applied over concrete block or gypsum board is used in wet areas and where cleanliness is a problem, such as in kitchens and toilet rooms. Occasionally walls with wood paneling or stone veneer are used for aesthetic effect. Doors are usually set in formed sheet-metal frames, although some wood frames are used. The doors themselves are usually made of solid timbers glue-laminated together and covered with thin decorative wood veneers; painted hollow sheet-metal doors are used for exterior doors and in areas of hard use.

Ceiling finishes. Ceiling finishes in these buildings create a sandwich space below the roof or floor slab above, which conceals projecting structural elements, recessed light fixtures, electrical wiring conduits, and air-handling ductwork. The ceiling must be accessible to change or maintain the service elements located above it, and the most common ceiling system is composed of wet felted mineral fibre panels, painted and perforated on one side for sound absorption. The removable panels are supported on a grid of formed sheet-metal tee bars or zee tracks, which are suspended by wires from the structure above. Where accessibility is not important and a smooth finish is desired, suspended gypsum board ceilings can be used.

Terrazzo

Floor finishes. Floor finishes in commercial and institutional uses make considerable use of synthetic-fibre carpeting and vinyl composition tile. In areas of higher traffic harder surfaces may be used—for example, cut stone tiles of marble or granite, ceramic tile applied with epoxy adhesive to the substrate, or terrazzo. Terrazzo is made in two ways, traditional and thin-set. In the traditional form a four-centimetre (1.5-inch) layer of cement and sand grout is poured over the substrate; a grid of metal divider strips to control shrinkage cracks is set on the hardened surface, and grout mix of coloured cement and marble chips is poured between the strips. After hardening, the surface is machine polished to expose the marble chips and metal dividers. Thin-set terrazzo is made by placing the metal strips and pouring the binder and marble chips directly onto the subfloor, without the underbed of cement and sand. It is generally possible only when epoxy resins are used in place of cement binders. Terrazzo is available in many colours, and it forms a hard, smooth, and durable surface that is easily cleaned.

Fire-suppression equipment

Building services. *Life-safety systems.* Most important in the hierarchy of interior elements are life-safety systems to protect and evacuate the building population in emergencies. These include life-threatening events, such as fire and smoke and earthquakes, and less critical ones, such as electric power failures. To deal with the threat of fire and smoke there is an array of fire-detection and fire-suppression systems. These include electronic heat and smoke detectors that can activate audible alarm devices to warn the building population and automatically notify local fire departments. For fire suppression hand-operated fire extinguishers must be provided, but many buildings have a separate piping system to provide water for fire fighting. If public water mains cannot provide adequate water pressure, an electric pump is included, and there is also a connection outside the building to attach portable fire truck pumps. The piping terminates in an array of sprinkler heads located throughout the building in the ceiling plane in a density ranging from eight to 18 square metres (90 to 200 square feet) per head. Typically there is always water in the pipes (a wet system), though dry systems are used in unheated buildings or where leakage might damage the contents. The head is opened to spray water by a fusible link made of metal that melts at a fairly low temperature when the air surrounding it is heated by a fire. Sprinkler systems have proved to be a highly reliable and effective means of fire suppression. Smoke can be as dangerous as fire to building occupants, and protective measures include the automatic shutdown of mechanical ventilating systems and the division of the building into smokeproof compartments to prevent the spread of smoke.

The evacuation of occupants in emergencies is accomplished by a system of protected exits leading to the exterior; all building areas must be within a specified travel distance of such an exit, varying from 30 to 90 metres (100 to 300 feet). For one-story buildings the exit usually consists simply of exterior doors, but for multistory buildings the exits are enclosed stairways that also lead to the exterior. The stairways have fire-rated enclosures and are often pressurized to exclude smoke; their width is determined by the maximum predicted number of occupants per floor. Travel paths to the exit must be clearly marked by illuminated directional exit signs, and battery-powered emergency lighting is required in the travel path and in the exit itself, in case of power failure. Some buildings of this type, such as hospitals, have large diesel- or natural gas-powered emergency electric generating systems that provide power and lighting for critical areas (such as operating rooms).

Another of the life-safety elements in these buildings is the fire-resistance requirements for building materials. These include the application of cementitious fireproofing or insulation to structural steel frames, the fire-resistive construction of the enclosures around exits, the flame-spread ratings of finish materials such as carpeting and wall coverings, and the use of such inherently fire-resistant materials as reinforced concrete and heavy timber. The fire-resistive ratings of various construction materials and assemblies are established by laboratory fire tests.

Vertical transportation. Vertical transportation systems in these low buildings include stairways, sometimes only those provided as life-safety exits but more often open, well-lighted ones as well. Where large numbers of people need to be moved vertically a short distance, escalators, or moving stairways, powered by electric motors are often provided. For moving smaller volumes of people and freight, hydraulic elevators are used; the cabs of these elevators are moved by a telescoping tubular piston underneath, which is raised and lowered by pumping oil in and out of it with an electric pump. Hydraulic elevators move slowly, but they are the least expensive type and are well suited for low buildings.

Plumbing. Plumbing systems for water supply and wastewater removal are very similar to those used in residential buildings, but the higher population densities of commercial, institutional, and industrial buildings require larger toilet rooms for public multiperson use. These often include pressure-valve water closets placed in partitioned cubicles and urinals in men's toilet rooms. Some fixtures in each toilet room must be carefully arranged for easy access by handicapped persons.

The internal drainage of large flat roofs introduces another piping system, similar to that for sanitary wastewater, to carry away storm water to separate underground storm sewers. Heavy rainstorms can introduce huge influxes of water into storm sewers, and sometimes this surge effect is tempered by the use of storm water retention ponds on the building site; runoff from the roof and paved areas is temporarily stored in these ponds while it flows into the sewer at a slower rate. Hospitals, laboratories, and factories have many other types of plumbing systems for various gases and liquids; these require special materials and construction. The sites of commercial, institutional, or industrial buildings may have underground networks of irrigation piping that terminate in flush sprinkler heads to water grass and plantings.

Environmental control systems. The atmosphere systems of industrial buildings are usually simple, involving only winter heating and possibly humidity control if the manufacturing process is sensitive to it. A commonly used element is the unit heater, in which an electric fan blows air through a coil heated by hot water, steam, electric resistance, or gas combustion and provides a directed supply of warm air where needed. Another system involves radiant heating using electric resistance coils backed by reflectors or continuous reflector-backed metal pipes that radiate heat from gas burned inside them. Ventilation in industrial buildings is sometimes done with operable windows but more often with unit ventilators, which penetrate walls or roofs and use electric fans to exhaust interior air that is replaced by air flowing in through operable louvres.

Rooftop atmospheric systems

Commercial and institutional low-rise buildings generally have fully controlled atmospheres with heating, cooling, and humidification. An economical method of providing this controlled atmosphere is with rooftop single or multizone package units. Each unit contains an electric fan to move conditioned air; heating elements, which can be gas or oil-fired or electric resistance coils; cooling coils, which use the compressive cooling cycle with compressor, cooling coils, and condensor coils to liberate heat; as well as a fresh-air intake and air exhaust. All of these elements are prefabricated in a rectangular enclosed unit that is simply set on the roof over an opening through which it is connected to the supply and exhaust ducts. The airflow over the heating and cooling elements can be partitioned to provide different conditioned airstreams to serve different zones of the building. The conditioned air is fed at a constant volume into treelike systems of insulated sheet-metal ductwork for transmission to the zones served. The conditioned air enters the occupied space through diffusers placed in the ceiling system and connected to the ducts by flexible spiral reinforced fabric tubes. Thermostats within the space sense temperatures and send signals by electricity or compressed airflow to the unit to adjust heating and cooling as required; relative humidity is held to a range of 20 to 40 percent. The return of air from the occupied space to the unit for reconditioning is sometimes done through a reverse tree of ductwork leading back to the unit, but more often in commercial buildings this is accomplished by placing the entire sandwich space between the ceiling and the structural deck above under negative pressure to make what is called a return-air plenum. The negative pressure is created by an opening into the plenum from the return side of the rooftop unit, and perforated openings or grills in the ceiling plane admit the return air from the occupied space. Return air can also be made to enter the plenum by passing over the lamps of fluorescent light fixtures; this permits the direct recovery of heat generated by the lamps, which can be recycled to the occupied space in winter.

The rooftop unit is best used in one-story buildings or smaller multistory ones. For larger multistory buildings, centralized atmosphere systems are used. These are built up of separate components, most of which are housed in mechanical equipment rooms or in penthouses at roof level. The components include fans for moving air, humidification devices, air-filtering devices, and refrigeration machines. Where large refrigeration machines are used, the condensor coils that liberate heat are no longer placed outside the building as in residential units or rooftop units but are located in a water jacket near the compressor. This water is circulated through a piping system to carry away the heat to a cooling tower outside the building where the water is sprayed into the atmosphere and partially evaporated to liberate heat, then recovered and returned at a lower temperature to the condensor coil jacket. Mechanical equipment rooms for atmosphere systems require a minimum of 5 percent of the floor space in a commercial building and can range up to 20 percent in hospitals and 40 percent in laboratory buildings; if the building is large, there can be more than one fan room with centralized refrigeration machines and cooling towers. The distribution of conditioned air in buildings with centralized atmosphere systems is usually done through an insulated ductwork tree, using the variable air volume (VAV) method. This method supplies conditioned air in variable amounts as required to maintain the desired temperature in occupied spaces; it results in considerable energy economies over constant volume air supply methods. Separate exhaust systems are used for areas generating heat and odours, such as kitchens, laboratories, and toilet rooms.

Electrical systems. Electrical systems in these buildings begin at a step-down transformer provided by the utility company and located within or very close to the building. The transformer reduces the standard line potential to two dual voltage systems, which then pass through master switches and electric meters to record the subscriber's usage. Each of the voltages provided serves a separate category of use; different levels are required for incandescent lights and small appliances, large appliances, ceiling-mounted non-incandescent lighting, and heavy machinery. Each voltage pair has a separate distribution system of wiring leading from the meters and master switches to circuit breaker panels, where it is further broken down into circuits similar to residential uses. Because high-voltage wiring is considered hazardous, the switches controlling overhead lighting use lower voltages, and each heavy machine has its own fused switch. From the circuit breaker panel, low-voltage power conduit and wiring is typically distributed through partitions and ceiling sandwich spaces, but, in large open areas of commercial buildings, there may be wireways embedded in the floor slab. These wireways can be either rectangular metal tubes inserted into the concrete slab before pouring or closed cells of formed steel deck; the wireways are tapped where desired to provide convenience outlets at floor level.

Lighting systems

Lighting in these buildings is predominantly fluorescent. Lamps range in size and wattage, and the available colours can range from warm white to cool white. Incandescent tungsten-filament lamps are used mostly for accent lighting, since their light-output efficiency is low. Mercury-vapour and metal halide-vapour lamps have the same efficiency as fluorescent lamps, but certain types may have longer operating lives. High-pressure sodium-vapour lamps have even higher efficiencies and are used in industrial applications; their marked orange colour and high intensity has limited their commercial and institutional use, however. Each of these types of lamp is used in a variety of fixtures to produce different lighting conditions. Incandescent lamps can be placed in translucent glass globes for diffuse effects, or in recessed ceiling-mounted fixtures with various types of reflectors to evenly light walls or floors. Fluorescent lamps are typically installed in recessed rectangular fixtures with clear prismatic lenses, but there are many other fixture types, including indirect cove lights and luminous ceilings with lamps placed above suspended plastic or metal eggcrate diffuser grids. Mercury-vapour and high-pressure sodium-vapour lamps are placed in simple reflectors in high-ceilinged industrial spaces, in pole-mounted light fixtures for outdoor applications on parking lots and roadways, and in indirect uplighting fixtures for commercial applications.

Mathematical models can accurately predict the performance of lighting in most applications. The zonal cavity method, which takes into account the lamps, fixtures, shape of room, and colours of room surfaces, is one example. The usual measure of light intensity is in footcandles on a horizontal surface, such as the floor of a room or a desk. The intensity ranges from 15 footcandles for a minimum ambient light level to 70 footcandles for an office or classroom and 100–200 footcandles for very precise visual tasks such as drafting; direct sunlight at noon, by comparison, is about 1,000 footcandles. In most of these buildings, the required lighting level is achieved with fixtures mounted at ceiling level; having all lighting at ceiling level allows flexibility in using building spaces. But the intensity of light varies inversely with the square of the distance from the source; thus, if a light fixture gives an intensity of 40 footcandles at a distance of one metre, it will produce an intensity of 10 footcandles at two metres. Therefore, considerable energy savings can be realized by having a minimal ambient light level (say 15 footcandles) produced by ceiling-mounted fixtures and providing task lighting close to work surfaces where higher intensities are needed. Daylighting from windows and skylights is also utilized in these buildings, and mathematical models have been developed that accurately predict its performance.

Communications systems

Communications systems are of growing significance and complexity in commercial, institutional, and industrial buildings. Thus communications wires for telephones, public-address systems, and computer data are free to take many paths through the building, including vertical risers, ceiling sandwich spaces, and wireways in floor slabs similar to those of electrical power wires. Where the density of wires rises to very high levels—for example, in computer rooms or where many small computer terminals are installed—raised floor systems are used. Removable floor panels are mounted on tubular metal frameworks resting on the structural floor slab, creating a plenum space to carry the necessary wiring.

A number of building systems are controlled by computers or microprocessors. In certain atmosphere systems both the interior sensors (such as thermostats) and the exterior weather sensors feed data to a computer that adjusts the system for minimal energy expenditure. Other examples include security, fire, and emergency alarm systems.

HIGH-RISE BUILDINGS

The high-rise building is generally defined as one that is taller than the maximum height which people are willing to walk up; it thus requires mechanical vertical transportation. This includes a rather limited range of building uses, primarily residential apartments, hotels, and office buildings, though occasionally including retail and educational facilities. A type that has appeared recently is the mixed-use building, which contains varying amounts of residential, office, hotel, or commercial space. High-rise buildings are among the largest buildings built, and their unit costs are relatively high; their commercial and office functions require a high degree of flexibility.

Foundations. The foundations of high-rise buildings support very heavy loads, but the systems developed for low-rise buildings are used, though enlarged in scale. These include concrete caisson columns bearing on rock or building on exposed rock itself. Bearing piles and floating foundations are also used.

Structural systems. *Wind loads.* The structural systems of tall buildings must carry vertical gravity loads, but lateral loads, such as those due to wind and earthquakes, are also a major consideration. Maximum 100-year-interval wind forces differ considerably with location; in the interiors of continents they are typically about 100 kilograms per square metre (20 pounds per square foot) at ground level. In coastal areas, where cyclonic storms such as hurricanes and typhoons occur, maximum forces are higher, ranging upward from about 250 kilograms per square metre (50 pounds per square foot). Wind forces also increase with building height to a constant or gradient value as the effect of ground friction diminishes. The maximum design wind forces in tall buildings are about 840 kilograms per square metre (170 pounds per square foot) in typhoon areas.

Effect of wind loads

The effect of wind forces on tall buildings is twofold. A tall building may be thought of as a cantilever beam with its fixed end at the ground; the pressure of the wind on the building causes it to bend with the maximum deflection at the top. In addition, the flow of wind past the building produces vortices near the corners on the leeward side; these vortices are unstable and every minute or so they break away downwind, alternating from one side to another. The change of pressure as a vortex breaks away imparts a sway, or periodic motion, to the building perpendicular to the direction of the wind. Thus, under wind forces there are several performance criteria that a high-rise structure must meet. The first is stability—the building must not topple over; second, the deflection, or sidesway at the top, must not exceed a maximum value (usually taken as 1/500 of the height) to avoid damage to brittle building elements such as partitions; and, third, the swaying motion due to vortex shedding must not be readily perceptible to the building occupants in the form of acceleration, usually stated as a fraction of gravity, or g. The threshhold of perception of lateral motion varies considerably with individuals; a small proportion of the population can sense 0.003 g or 0.004 g. The recommendation for motion perception is to limit acceleration to 0.010 g for wind forces that would recur in 10-year intervals. The fourth criterion involves the natural period of the building structure. This is the vibration period at which the swaying cantilever motions of the building naturally reinforce and enhance each other and could become large enough to damage the building or even cause it to collapse. The natural period of the building should be less than one minute, which is the period of vibration due to the shedding of wind vortexes.

Earthquake loads. Earthquake or seismic forces, unlike wind forces, are generally confined to relatively small areas, primarily along the edges of the slowly moving continental plates that form the Earth's crust. When abrupt movements of the edges of these plates occur, the energy released propagates waves through the crust; this wave motion of the Earth is imparted to buildings resting on it. Timber frame buildings are light and flexible and are usually little damaged by earthquakes; masonry buildings are heavy and brittle and are susceptible to severe damage. Continuous frames of steel or reinforced concrete fall between these extremes in their seismic response, and they can be designed to survive with relatively little damage.

Designing for lateral acceleration

In two major earthquakes involving large numbers of high-rise buildings, in Los Angeles in 1971 and Mexico City in 1985, lateral accelerations due to ground motions in a number of tall buildings were measured with accelerometers and were found to be of the order of 0.100 to 0.200 g. In Los Angeles, where the period of the seismic waves was less than one second, most steel-frame high rises performed well with relatively little damage; continuous concrete frames also generally performed well, but there was considerable cracking of concrete, which was later repaired by the injection of epoxy adhesive. In Mexico City, however, the period of the seismic waves was quite long, on the order of a few seconds. This approached the natural frequency of many tall structures, inducing large sidesway motions that led to their collapse. Based on this experience, determination of the seismic performance criteria of buildings involves the lateral resistance of forces of 0.100 to 0.200 g and consideration of the natural period of the building in relation to the period of seismic waves that can be expected in the locality. Another important factor is the ductility of the structure, the flexibility that allows it to move and absorb the energy of the seismic forces without serious damage. The design of buildings for seismic forces remains a complex subject, however, and there are many other important criteria involved.

Classification of structural systems. The types of structures used for high-rise buildings must meet the lateral load performance criteria outlined above, and they must be reasonably efficient in the use of material and of reasonable cost. The most efficient high-rise structure would meet the lateral load criteria using no more material than would be required for carrying the building gravity load alone; in other words, it would have no premium

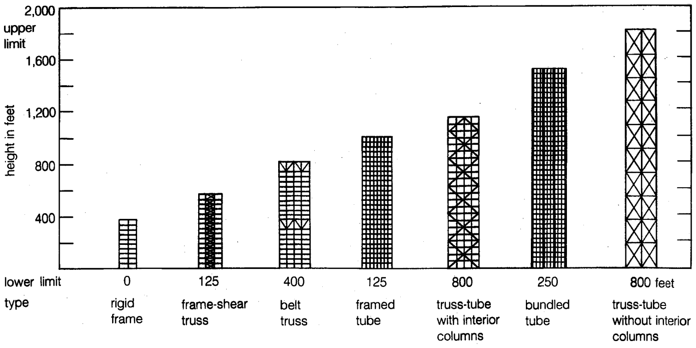


Figure 12: Structural systems for tall buildings.

From A. Swenson/P.-C. Chang, *Architectural Education at IIT, 1938–1978* (1980), Illinois Institute of Technology Press

for height. This economic criterion of "no premium for height" has led to a classification of high-rise structures, each of which has only a small premium for a particular range of height (Figure 12).

High-rise structures begin at the lowest range with the rigid frame in both steel and concrete. Some or all of the joints between the beams and columns are rigidly joined together by welding the steel or pouring the concrete in situ, and lateral resistance is provided by the rigid joints; this system can rise about 90 metres (300 feet) with little premium. The next type is the rigid frame with a vertical shear truss in steel or a shear wall in concrete to provide greater lateral rigidity; it has a range of 38 to 150 metres (125 to 500 feet). The framed tube structure in both steel and concrete brings more gravity load and more structural material to closely spaced columns at the building's perimeter, again increasing lateral rigidity; this type is reasonably efficient from 38 to 300 metres (125 to 1,000 feet) in height. The trussed tube with interior columns, which can also be executed in both steel and concrete, introduces diagonal bracing on all sides of the building's perimeter. The bracing also carries gravity loads and further raises the lateral rigidity, making this a low-premium structure for the region of 240 to 360 metres (800 to 1,200 feet). The bundled tube, which consists of a number of framed tubes joined together for even greater lateral rigidity, begins to be practical at about 75 metres (250 feet). It was the form of the steel structure used for the Sears Tower in Chicago. Beyond this height there is another system that appears to have a low premium: the superframe. In this structure much of the building's gravity load, and therefore its material, is brought to a diagonally braced superframe tube at the perimeter by interior transfer trusses of various configurations. No true superframes have yet been built.

Enclosure systems. The enclosure systems for high-rise buildings are usually curtain walls similar to those of low-rise buildings. The higher wind pressures and the effects of vortex shedding, however, require thicker glazing and more attention to sealants. The larger extent of enclosed surfaces also requires consideration of thermal movements, and wind- and seismic-induced movements must be accommodated. Window washing in large buildings with fixed glass is another concern, and curtain walls must provide fixed vertical tracks or other attachments for window-washing platforms. Interior finishes in high-rise buildings closely resemble those used in low-rise structures.

Building services. *Life-safety systems.* Life-safety systems are similar to those in low-rise buildings, with stairways serving as vertical emergency exits; in case of fire all elevators are automatically shut down to prevent the possibility of people becoming trapped in them. Emergency generator systems are provided to permit the operation of one elevator at a time to rescue people trapped in them by a power failure. Generators also serve other vital building functions such as emergency lighting and fire pumps. Fire-suppression systems often include sprinklers, but, if none are required by building codes, a separate piping system is provided with electric pumps to maintain pressure and to bring water to fire-hose cabinets throughout the building. There are also exterior connections at street level for portable fire-truck pumps. The fire hoses are so placed that every room is accessible; the hoses are intended primarily for professional fire fighters but may also be used by the building occupants.

Vertical transportation. Vertical transportation systems are of vital importance in high-rise buildings. Escalators are used on lower floors for moving high volumes of people over short distances. A few retail or educational buildings have escalators for up to 10 stories. The principal means of vertical transport in tall buildings is the roped elevator. It moves by a direct current electric motor, which raises and lowers the cab in a shaft with wire ropes running over a series of sheaves at the motor and the cab itself; the ropes terminate in a sliding counterweight that moves up and down the same shaft as the cab, reducing the energy required to move the elevator. Each elevator cab is also engaged by a set of vertical guide tracks and has a flexible electric cable connected to it to power lighting and doors and to transmit control signals. Passenger elevators range in capacity from 910 to 2,275 kilograms (2,000 to 5,000 pounds) and run at speeds from 90 to 510 metres per minute; freight elevators hold up to 4,500 kilograms (10,000 pounds). The speed of elevators is apparently limited to the current value of 510 metres per minute by the acceleration passengers can accept and the rate of change of air pressure with height, which at this speed begins to cause eardrum discomfort.

Roped elevators

Elevator movements are often controlled by a computer that responds to signals from call buttons on each floor and from floor-request buttons in each cab. The number of elevators in a building is determined by the peak number of people to be moved in a five-minute period, usually in the early morning; for example, in an office building this is often set at 13 percent of occupancy. The average waiting time for an elevator between pressing the call button and arrival must be less than 30 seconds in an office building and less than 60 seconds in an apartment building. The elevators are usually arranged in groups or banks ranging from one to 10 elevators serving a zone of floors, with no more than five elevators in a row to permit quick access by passengers. In a few very tall buildings the sky lobby system is used to save elevator-shaft space. The building is divided vertically into subbuildings, each with its own sky lobby floor. From the ground floor large express elevators carry passengers to the sky lobby floors, where they transfer to local elevator banks that take them to the individual floors within the subbuildings.

Plumbing. Plumbing systems in tall buildings are similar to those of low-rise buildings, but the domestic water-supply systems require electric pumps and tanks to maintain pressure. If the building is very tall, it may require the system to be divided into zones, each with its own pump and tank.

Environmental control systems. The atmosphere systems in high-rise office buildings are similar to those of low-rise, with conditioned air distributed by a ductwork tree using the VAV system and return air removed through ceiling plenums. The placement of air-handling equipment can be done in two ways. One uses centralized fans placed about every 20 floors, with air moved vertically through trunk ducts to and from each floor; the other uses floor-by-floor fan rooms to provide air separately for each floor. There is usually a central refrigeration plant for the entire building connected with cooling towers on the roof to liberate heat. The central refrigeration machines produce chilled water, which is circulated by electric pumps in a piping system to the air-handling fans in order to cool incoming air as required. Incoming air is heated in winter either by piping coils through which hot water is circulated by pumps and piping from a central boiler, or by electric resistance coils in the air-handling units. In residential high-rise buildings cooling is typically provided by window air-conditioning units, and heating by hot-water or electric resistance radiant systems. There is limited use of centralized cooling, in which chilled water from a central refrigeration plant is circulated to fan-coil units near the building perimeter; a small electric fan within the unit circulates the air of the room over the chilled water coil to absorb heat.

Electrical systems. Electrical systems for high-rise buildings are also very similar to low-rise types. The major difference is that, if the building is exceptionally tall, the utility company may bring its high-voltage lines inside the building to a number of step-down transformers located in mechanical equipment spaces. From each step-down transformer the distribution of electricity is similar to that of a smaller building.

LONG-SPAN BUILDINGS

Types of long-span buildings

Long-span buildings create unobstructed, column-free spaces greater than 30 metres (100 feet) for a variety of functions. These include activities where visibility is important for large audiences (auditoriums and covered stadiums), where flexibility is important (exhibition halls and certain types of manufacturing facility), and where large movable objects are housed (aircraft hangars). In the late 20th century, durable upper limits of span have been established for these types: the largest covered stadium has a span of 204 metres (670 feet), the largest exhibition hall has a span of 216 metres (710 feet), and the largest commercial fixed-wing aircraft has a wingspread of 66.7 metres (222 feet) and a length of 69.4 metres (228 feet), requiring a 75–80-metre- (250–266-foot-) span hangar. In these buildings the structural system needed to achieve these spans is a major concern.

Structural systems. *Structural types.* Structural systems for long-span buildings can be classified into two groups: those subject to bending, which have both tensile and compressive forces, and funicular structures, which experience either pure tension or pure compression. Since bridges are a common type of long-span structure, there has been an interplay of development between bridges and long-span buildings. Bending structures include the girder, the two-way grid, the truss, the two-way truss, and the space truss. They have varying optimum depth-to-span ratios ranging from 1 : 5 to 1 : 15 for the one-way truss to 1 : 35 to 1 : 40 for the space truss. The funicular structures include the parabolic arch, tunnel vault, and dome, which act in pure compression and which have a rise-to-span ratio of 1 : 10 to 1 : 2, and the cable-stayed roof, the bicycle wheel, and warped tension surfaces, which act in pure tension. Within these general forms of long-span structure, the materials used and labour required for assembly are an important constraint along with other economic factors (Figure 13).

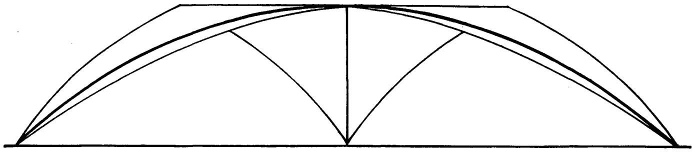
Timber structures. Glue-laminated timber can be used as a long-span material. It can be prefabricated using metal connectors into trusses that span up to 45 metres (150 feet). Its most economical forms, however, are the pure compression shapes of the multiple-arch vault, with spans up to 93 metres (305 feet), and ribbed domes, with spans up to 107 metres (350 feet). These are often used as industrial storage buildings for materials such as alumina, salt, and potash that would corrode steel or concrete. Such timber structures are usually found only near forested areas; transportation of timber to other areas increases its cost.

Steel structures. Steel is the major material for long-span structures. Bending structures originally developed for bridges, such as plate girders and trusses, are used in long-span buildings. Plate girders are welded from steel plates to make I beams that are deeper than the standard rolled shapes and that can span up to 60 metres (200 feet); however, they are not very efficient in their use of material. Trusses are hollowed-out beams in which the stresses are channeled into slender linear members made of rolled shapes that are joined by welding or bolting into stable triangular configurations. The members of trusses act either in pure compression or pure tension: in the top and bottom horizontal members the forces are greatest at the centre of the span, and in the verticals and diagonals they are greatest at the supports. Trusses are highly efficient in bending and have been made up to 190 metres (623 feet) in span. Two-way grids can be made of either plate girders or trusses to span square spaces up to 91 metres (300 feet) in size; these two-way structures are more efficient but more expensive to build.

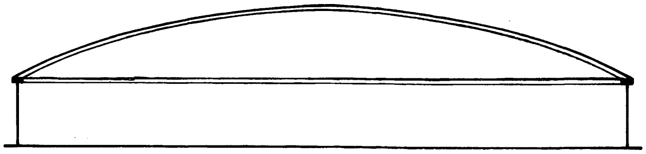
Steel funicular structures

The highly efficient funicular forms are used for the longest spans. Vaults made of rows of parabolic arches, usually in truss form for greater rigidity, have been used for spans of up to 98.5 metres (323 feet). Steel truss domes, particularly the Schwedler triangulated dome, have been the choice for several large covered stadiums, with the greatest span being 204.2 metres (669 feet). Cable-stayed roof construction is another structural system derived from bridge building. A flat roof structure in bending is supported from above by steel cables radiating downward from masts that rise above roof level; spans of up to 72 metres (236 feet) have been built. Another funicular form is the bicycle-wheel roof, where two layers of radiating tension cables separated by small compression struts connect a small inner tension ring to the outer compression ring, which is in turn supported by columns.

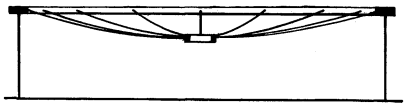
Tension-cable networks use a mesh of cables stretched from masts or continuous ribs to form a taut surface of negative curvature, such as a saddle or trumpet shape; the network of cables can be replaced by synthetic fabrics to form the tension surface. Another fabric structure using tension cables is the air-supported membrane. A network of cables is attached by continuous seams to the fabric, and the assembly of cables and fabric is supported by a compression ring at the edge. The air pressure within the building is increased slightly to resist exterior wind pressure. The increase can be as slight as 1.5 percent of atmospheric pressure, and it is possible to maintain this even in large buildings with relatively small compressors. The cables stiffen the fabric against flutter under uneven wind pressure and support it in case of accidental deflation.



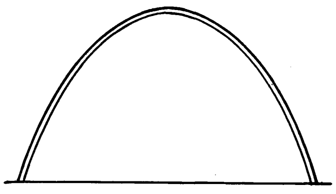
concrete shell 216 m (720 ft)



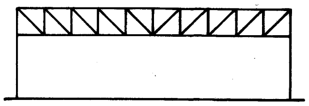
steel dome 204 m (678 ft)



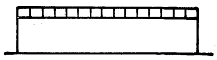
concrete cable dish 126 m (420 ft)



steel truss vault 98.5 m (328 ft)



steel truss 90 m (300 ft)



steel plate girder 60 m (200 ft)

Figure 13: Long-span structures.

Concrete shell vaults

Concrete structures. Reinforced concrete, because of its inherent strength in compression, is primarily used for long spans in funicular compression forms, including vaults, shells, and domes. Thin parabolic shell vaults stiffened with ribs have been built with spans up to about 90 metres (300 feet). More complex forms of concrete shells have been made, including hyperbolic paraboloids, or saddle shapes, and intersecting parabolic vaults. An example of the latter is the CNIT Exhibition Hall in Paris, which consists of six intersecting double-shell parabolic vaults built to span a triangular space 216 metres (708 feet) on a side with supports only at the apexes of the triangle. Reinforced concrete domes, which are usually also of parabolic section, are built either in ribbed form or as thin shells. The maximum span of these domes is about 200 metres (660 feet).

Another funicular form used in concrete, though it is really a composite structure, is the inverted dome, or dish. As in the steel bicycle wheel, a concrete compression ring resting on columns at the perimeter of the structure supports radial steel cables that run inward and downward to a small steel tension ring at the centre, forming the dish shape. The cable network is stiffened against wind forces by encasing it in a poured concrete dish; structures of this type have been built with spans of up to 126 metres (420 feet).

Factors in the built environment. *Acoustics.* Long-span auditoriums involve considerations in acoustics: audiences wish to hear speakers clearly and to hear music with appropriate tonality. Unfortunately, acoustic requirements for speech quality often conflict with those for music, and it is difficult to design an auditorium that is satisfactory for both. The best single measure of acoustic performance for auditoriums is the reverberation time, which is directly proportional to the volume of the hall and inversely proportional to the amount of sound absorbency within it, including wall and ceiling surfaces and the audience itself. Measured in the sound range of 500–1,000 hertz, rooms with short reverberation times of one to 1.5 seconds are good for the intelligibility of speech, while longer reverberation times of 1.5 to 2.5 seconds add richness of tone to musical performances. Thus, adding sound-absorbent material to a hall improves it for speech but detracts from its musical qualities. People are excellent sound absorbers, and thus the audience has a distinct impact on auditorium acoustics; to keep this effect constant with varying audience size, auditorium seats are usually upholstered to serve as surrogate spectators of the same sound absorbency. Curved surfaces, which tend to focus sound, are either avoided in auditoriums or covered with sound-absorbent material. Electronic sound-amplification systems can be used to assist speakers in large halls but generally are not satisfactory for music. Other long-span buildings, such as covered stadiums and exhibition halls, receive only minor acoustical treatment.

Environmental control systems. Atmosphere systems in long-span buildings must handle the considerable heat and odour generation from population densities of less than one square metre (11 square feet) per person. Air must be moved fairly rapidly through the population zone to maintain an acceptable air-change rate.

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(A.Sw./P.-C.C.)

Bulgaria

The People's Republic of Bulgaria (Narodna Republika Bŭlgariya) occupies the eastern portion of the Balkan Peninsula in southeastern Europe. The country is crossed by historically important routes from northern and eastern Europe to the Mediterranean basin and from western and central Europe to the Middle East. Bulgaria is remarkable for its variety of scenery, its rugged mountains and modern Black Sea resorts alike attracting tourists from other countries. Although the Ottoman influence remains visible, Bulgaria's chief historical, cultural, and political ties connect it with the Soviet Union. Long-term alliances with other eastern European countries became uncertain in late 1989 as a sweeping democratic-reform movement dissolved the Soviet bloc. In January 1990 the Bulgarian parliament revoked the Communist Party's constitutionally guaranteed monopoly on political power.

A nation of some 43,000 square miles (111,000 square kilometres), Bulgaria is bounded by Romania on the north, most of the border being marked by the lower Danube River. The Black Sea lies to the east, Turkey and Greece to the south, and Yugoslavia to the west. The capital of the country is Sofia, which lies in a mountainous basin in the west, near the geographic centre of the whole Balkan region.

This article is divided into the following sections:

Physical and human geography

THE LAND

Topographic variety

Relief. Within a relatively small compass, the Bulgarian landscape exhibits striking topographic variety. Open expanses of lowland alternate with broken mountain country, cut by deep river gorges and harbouring upland basins such as that in which Sofia lies. Three basic structural and physiographic divisions run east–west.

The Danubian Plain. All but a short section of the northern frontier of Bulgaria is marked by the lower Danube, the abrupt and often steep banks on the Bulgarian side contrasting with the swamps and lagoons of the Romanian side. A fertile, hilly plain extends southward from the Danube to the foothills of the Balkan Mountains. The average elevation of the region is 584 feet (178 metres), and it covers some 12,200 square miles (31,500 square kilometres). Several rivers cross the plain, flowing northward from the Balkans to join the Danube.

The Balkan Mountains. The Balkan Mountains border the Danubian Plain on the south, their rounded summits having an average height of 2,368 feet and rising to 7,795 feet (2,376 metres) in Botev, the highest peak. Parallel to this principal chain (its Bulgarian name, Stara Planina, means "Old Mountains," "old" signifying its greater extent compared to the adjacent ranges) lies a transitional region of complex relief. Block faulting—the raising or lowering of great structural segments along regular lines of crustal weakness—has produced there the Sredna Mountains (Sredna Gora), the Vitosha Mountains, the Lisa Mountains, a number of sheltered structural basins, and the Upper Thracian and Tundzha lowlands.

The Rila–Rhodope massif. Another mountain mass covers southern Bulgaria. This includes the alpine Rhodope Mountains (Rodopi in Bulgarian, Rodhópis in Greek), which rise to 7,188 feet (2,191 metres) in Golyam Perelik Peak; the Rila Mountains, rising to Musala Peak, at 9,596 feet (2,925 metres) the highest in the country and indeed in the whole Balkan Peninsula; the Pirin Mountains, with Vikhren Peak reaching 9,560 feet (2,914 metres); and a frontier mountain range known as the Belasitsa.

These majestic ranges discharge meltwater from montane snowfields throughout the summer, and their sharp outlines, pine-clad slopes, and, in the Rila and Pirin ranges, several hundred lakes of glacial origin combine to form some of the most beautiful of Bulgarian landscapes.

The Black Sea coast. Trending north–south at the eastern fringe of the other regions is the narrow Black Sea coastal region. The coast has few bays (exceptions being the fine harbours of Varna and Burgas) but does have extensive stretches of sandy beach that have led to the growth of a number of picturesque seaside resorts.

Rivers

Drainage. Bulgaria has a complex drainage pattern characterized, with the notable exception of the Danube, by relatively short rivers. The major rivers are the Maritsa (Marica), the Iskŭr, the Struma, the Arda, the Tundzha, and the Yantra. Overall, more than half of the runoff drains to the Black Sea, and the rest drains to the Aegean Sea. Bulgaria's numerous lakes may be coastal (such as the large lakes around Varna and Burgas, both on the Black Sea), glacial (such as those in the southern mountains), structural, or karst in origin. The country has some 500 mineral springs, half of which are warm or hot (reaching 217° F [103° C] at Sapareva Banya, in the west). Numerous dams have been constructed in the mountains.

Soils. The varied Bulgarian natural environment has produced about 20 soil types and subtypes, which may be grouped into three main regions. Northern Bulgaria is characterized by the fertile black-earth soils known as chernozems and also by gray soils of forest origin. Southern Bulgaria has forest soils with acidic (cinnamonic) traces—by far the most extensive single category—as well as the modified chernozems known as chernozem-smolnitzas (a dark-coloured zonal soil with a deep and rich humus horizon). The rugged high mountain regions have brown forest, dark mountain forest, and mountain meadow soils.

Precipitation

Climate. The greater part of Bulgaria has a moderate continental climate, which is tempered by Mediterranean influences in the south. The average annual temperature is 51° F (10.5° C), but this conceals a wide variation; temperatures as low as −37° F (−38° C) and as high as 113° F (45° C) have been recorded. Mean annual precipitation ranges from about 18 inches (450 millimetres) in the northeast to more than 47 inches in the highest mountains. The lowlands receive snowfall from mid-October to mid-May, with an annual average of 25–30 days of snow cover. Hailstorms occur between May and August.

Plant and animal life. The relatively large number of Bulgarian plant and animal species reflects the country's location adjoining several of the great Eurasian biogeographic zones. During the Ice Age, life in the region was not destroyed by advancing glaciers as occurred in much of Europe but was actually enriched by the immigration of species from the north, some of which still survive. Influences from the steppes of western Asia also penetrated the region at that time. Nonetheless, the greater part of the plant and animal life is central European, mixed with a type that blends Arctic and alpine characteristics in the high mountains. Steppe species are most characteristic of the northeast and southeast, while the south is rich in sub-Mediterranean and Mediterranean species. The Bulgarian government has introduced a number of conservation measures, including steps to protect soil, water, and air from pollution and to establish protected areas of outstanding interest to naturalists.

Settlement patterns. The natural environment of Bulgaria falls into three basic regions: North Bulgaria, including the Danubian Plain and the Balkan Mountains; South Bulgaria, including the Rila–Rhodope massif; and a transitional area between them. Each of these traditional regions over the centuries has been subjected to the actions of humans, whether in the remoter mountains, where national culture was cherished and the seeds of nationalism were sown during the long Ottoman domination, or in the Danubian Plain, where agriculture has been practiced from ancient times. Bulgarian settlements have been officially classified into more than 200 urban areas and 5,000 villages, the latter including hundreds of small hamlets, clusters of farmsteads, and, deep in the mountains, a handful of historic monasteries.

Urban patterns. Many Bulgarian towns have roots in the Middle Ages and in remotest antiquity, although new, modern settlements are being created each year; the urban population overtook the rural for the first time in 1969. Each new or modernized town is required to have its own development plan, and the neighbourhood unit is a characteristic feature of the planning process. Despite the pressure of urban population growth, many Bulgarian towns preserve their ancient charm and are rich in cultural monuments; in the remoter areas, they offer a slower pace of life than can be found in the cities.

Major cities

Sofia, the capital, is the largest city and dominates the national economic and cultural life. Plovdiv, in the south-central region, enjoys a scenic location on the Maritsa River and is another major industrial and cultural centre, where an international trade fair is held annually. Varna is a centre for industry, transport, and tourism on the shores of the Black Sea. The nearby seaside resort of Zlatni Pyassŭtsi ("Golden Sands") attracts an international tourist trade. Burgas is Bulgaria's largest Black Sea port and a major industrial, cultural, and resort centre. Ruse, on the Danube in the north, is the largest Bulgarian river port; there the Friendship Bridge leads to the Romanian city of Giurgiu. Stara Zagora lies on the southern flanks of the central Sredna Mountains and is notable for its archaeological and architectural remains.

Rural patterns. The numerous contemporary Bulgarian villages are either clustered or scattered. Since World War II they have undergone a transformation from the sleepy, backward, and poverty-stricken settlements that typified much of the region for centuries. Almost all of the rural population live in villages supplied with water and electricity; three-quarters of the houses are recent constructions, replacing the older lath and plaster structures. Most of the villages have paved and asphalted streets. Processing plants have been built in many villages, so that rural areas are increasingly industrialized.

THE PEOPLE

The origins of the Bulgarians

Ethnic, religious, and linguistic background. Ethnically, the population is fairly homogeneous, Bulgarians making up about 85 percent of the total. The Slav tribes that settled in the eastern part of the Balkan Peninsula in the 6th and 7th centuries AD, thereby assimilating the local Thracian tribes, formed a basic ethnic group. The group known as the proto-Bulgarians, who, together with the Slavs, formed the Slav-Bulgarian state in 681, formed another component. With the gradual obliteration of fragmented Slav tribes and also of the ethnic differences in the territory of the first Bulgarian state (covering Moesia, Thrace, and Macedonia) in the 9th and 10th centuries, proto-Bulgarians and Slavs consolidated into a unified Slav people who thence retained the name of Bulgarians. This national unity, present in embryonic form during the long Ottoman domination, flowered in the independence struggles of the 19th century.

The Turks, Bulgaria's largest minority, live in some regions of the northeast and in the eastern Rhodope. Gypsies and Macedonians are two other sizable minorities (though the government does not consider Macedonians as such, regarding them as ethnically Bulgarian), and there are a few thousand Armenians, Russians, and Greeks (mostly in the towns) and Romanians and Tatars (mostly in the villages).

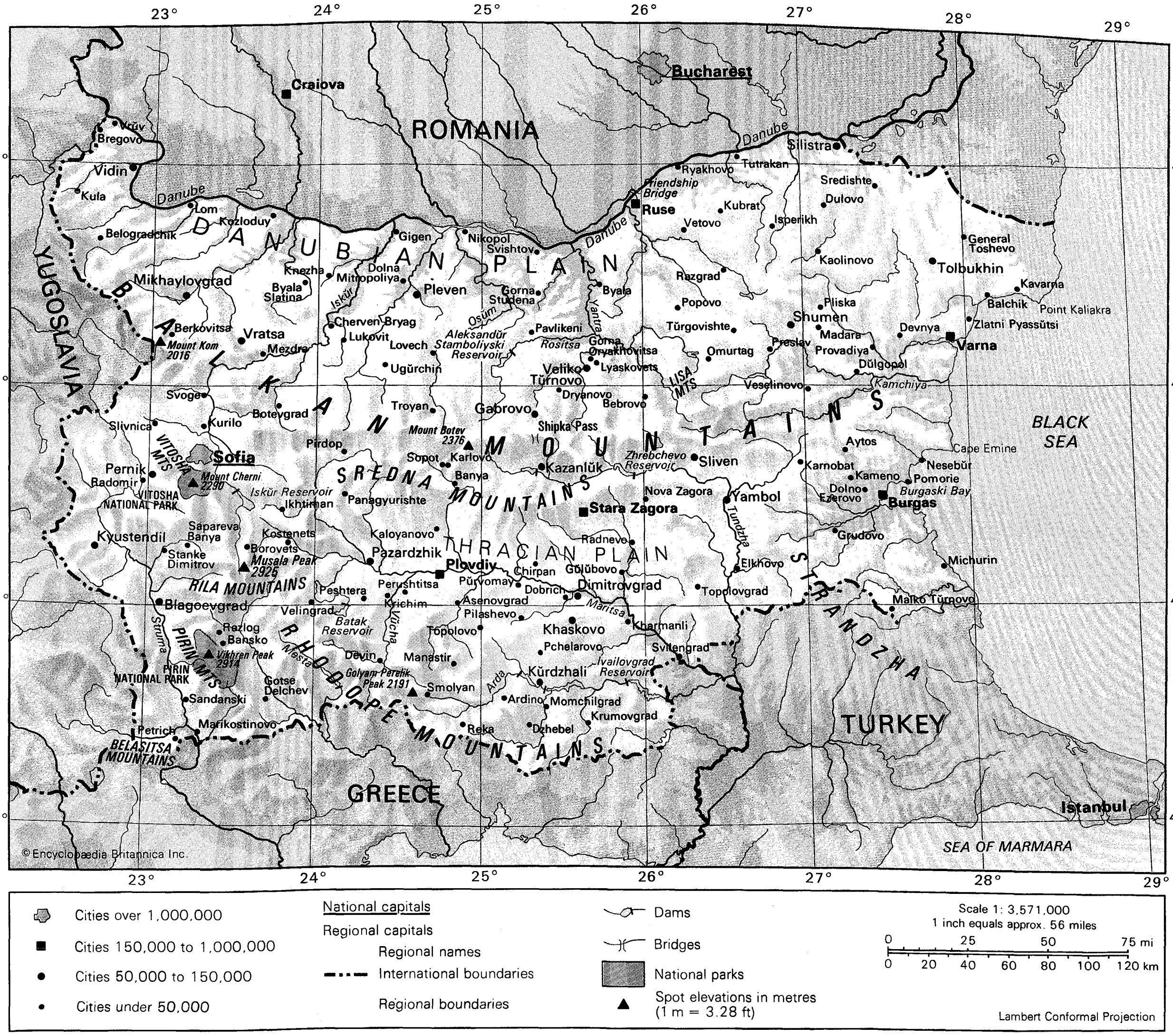
The Bulgarian language belongs to the South Slavic group, along with Serbo-Croatian, Slovene, and Macedonian (the last, however, considered to be a dialect of Bulgarian by the government). A number of dialects remain in the common speech.

Following the reforms of 1989–90, the state continued to encourage atheism, but religious activity was increasingly tolerated. The majority of religious Bulgarians are adherents of the Bulgarian Orthodox Church, which pressed the new leadership for greater autonomy and restoration of monastic lands. Minority religious groups include Muslims, Jews, Bulgarian Catholics, Protestants, and Gregorian Armenians.

Demographic trends. *Birth and death rates.* As a result of social and economic changes after World War II, notably the introduction of free medical care and the improvement of working conditions, Bulgaria's death rate dropped significantly, but it began to rise again in the 1970s as the proportion of older people in the population rose. The birth rate has dropped, as has the infant mortality rate.

Emigration and immigration. Emigration since World War II has mostly affected non-Bulgarians. About half a million Turks left the country, 155,000 having been expelled in 1949–51 and a further 300,000 or more having emigrated in 1989. Almost all Czechs and Slovaks returned to their homelands, as did large numbers of Russians and Armenians (to the Soviet Union) and Jews (to Israel). About 35,000 others, mostly Bulgarians, returned from North America and from European countries. Internally, the movement of population has been from rural areas to larger towns and cities. From 1946 to 1988, for example, the population of Sofia trebled; the populations of Varna and Plovdiv increased more than sixfold; and Ruse's population increased more than fivefold.

Population distribution. Bulgaria's geographic variety is reflected in the distribution of its population. The most densely populated areas are the Danubian Plain, the Upper Thracian Basin, the Burgas Plain, and the intermontane basins of southwestern Bulgaria. Areas of lowest density are the east and southeast of the country, the Strandzha and Dobruja regions, and the higher mountain areas.



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Urbanization continues to have an effect on the demographic structure; a large segment of the urban population is of a young working—and childbearing—age, leading to natural growth of the towns. Because relatively more older adults remain in the villages, the birth rate there continues to be lower and the death rate higher. These effects thus amplify the shift of population from rural areas to urban centres.

THE ECONOMY

Until the reform movement of the late 1980s, the Bulgarian economy was based solely on the premise that all means of production should be owned by the state. The National Assembly elected in 1990, while dominated by the Bulgarian Socialist Party (formerly called Bulgarian Communist Party), was to draft a new constitution incorporating reforms.

The state's role in agriculture and industry. Beginning in 1990, farmers were paid bonuses for each head of cattle or sheep exceeding the previous year's total; reduction of herds or flocks resulted in fines. Hard-currency exchange rights gave dairy farmers an incentive to maintain high milk-production levels. Supplies of quality wool and hides were to be paid for in hard currency. Farmers in upland and mountainous regions also received incentives. In addition, the government lifted restrictions on private farming and made available loans for the establishment of small farms and food-processing facilities.

Turnover tax

Taxation. The main sources of revenue are the turnover tax and deductions made from the profits produced by public enterprises. The turnover tax, a form of value-added tax, is based on a fixed rate and goes immediately into the budget after the sale of products by state enterprises. In this way the state budget is ensured a regular and uniform source of revenue to finance the undertakings called for in the economic plan. The turnover tax is dependent on the size, variety, and sale of manufactured products; ultimately it is passed on to the consumer. The profit deduction tax from state enterprises, unlike the turnover tax, is not at a fixed rate. It comes from each enterprise's net income after deduction of the turnover tax. The profit shown by an enterprise is the difference between income and maintenance expenses.

Budget. The national budget finances most capital investments, enterprises under direct central management, and a number of social and cultural needs (*e.g.,* higher educational institutions). It also covers defense and the central government departments. People's Council budgets—those of the district, municipality, and village People's Councils—maintain a variety of educational, health, and cultural institutions, urban development and planning, and various other services. The state social insurance budget covers expenditure for matters such as employees' pensions, temporary incapacity to work, maternity leave, maintenance of rest homes, and family allowances.

The national budget allots about half of its revenues to developing the national economy, a large portion of this going to development of heavy industry. It also provides funds for capital investment in industrial enterprises and provides them with turnover funds. Large credits are allocated to agricultural development and the supply of farm equipment. Budgetary credits are also used to finance long-term investment projects (*e.g.,* power generation, forestry, ferrous metallurgy, and land improvement and water conservation projects). About a third of all investment comes from state budgetary credits, and two-thirds is covered by local enterprises' funds and bank credits. About one-fourth of the total budgetary expenditure goes to fund social services.

The role of trade unions. Bulgaria has some 17,000 local trade union organizations made up of 111,000 separate trade union subgroups. Only an insignificant portion of the country's work force does not belong to a trade union. Until it was reconstituted as the Confederation of Independent Bulgarian Trade Unions (S'uz na Nezavisemite B'lgarski Profs'uze) in late 1989, all trade unions belonged to the Central Council of Trade Unions (Tsentralen Sŭvet na Profesionalnite Sŭyuzi), founded in 1944 and allied with the Bulgarian Communist Party. Also in 1989, Podkrepa ("Support"), a leading independent trade union, along with both established and newly formed opposition parties, formed the Union of Democratic Forces (S'uz na Demokraticheskite Sili) in order to press for governmental reform.

Coals

Resources. Bulgaria is relatively well endowed with a variety of both metallic and nonmetallic minerals. Geologic exploration has identified about 40 coal basins, which together contain an estimated 4.1 billion tons of proven recoverable reserves. Of the reserves, more than 99 percent is lignite. The main mining areas are in the Pernik basin southwest of Sofia, the Maritsa basin south of Stara Zagora, the Maritsa basin at Dimitrovgrad in the south, and Lom on the Danube. Lignite and brown coal fuel the country's thermal power stations and are used as fuel and as raw material for many of its industries. Although deposits of bituminous and anthracite coal have been almost exhausted in Bulgaria, other promising deposits of black coking coal have been found in the northeast, in the Dobruja region.

Deposits of iron ore are estimated at 317 million tons; one of the largest reserves is at Kremikovtsi near Sofia, the site of the country's largest metallurgical plant. Smaller quantities of iron ore are mined in the northwest (Mikhaylovgrad), in the central region (Troyan), and in the southeast (Yambol). There are significant deposits of nonferrous ores (copper, lead, and zinc) in the Rhodope Mountains, the Balkan Mountains, and the Sredna Mountains. Bulgaria is also rich in less valuable minerals, including rock salt, gypsum, limestone, dolomite, kaolin, asbestos, perlite, feldspar, fluorite, and barite.

Bulgaria has only small deposits of oil and natural gas; mineralogists have begun offshore exploration of the Black Sea, which is believed to be rich in coal, oil, natural gas, and other minerals.

Bulgaria gets about one-fourth of its electrical power from nuclear plants, and hydroelectric plants supply one-tenth.

Industry. Before World War II, industries were of minor importance in Bulgaria. Under the Socialist system, industrialization became one of the principal aims of economic policy, with particular emphasis on basic industries such as electric power, ferrous and nonferrous metallurgy, and chemicals. Central planning of management, production, and investment channeled into industry a large portion of the national resources.

Postwar industrialization

Before the war, shipbuilding at Varna and foundries at Sofia, Plovdiv, Ruse, and Pernik were the most important metallurgical industries. Those developed after the war include iron and steel works at Pernik, based upon local brown coal and iron ore from the Sofia district; a large steel project at Kremikovtsi; a lead and zinc works at Kŭrdzhali; and a copper and sulfuric acid plant at Pirdop. A chemical industry has developed at Dimitrovgrad, and chemical plants were also built at Stara Zagora, Vratsa, Devnya, and Vidin, and a petrochemical plant at Burgas. The biotechnology sector is increasingly important in the economy.

Machine building is a priority of the national economy. Its relative share of industrial production has jumped dramatically. Machine building and metal processing are widely dispersed throughout the country; the largest plants are located in Sofia, Varna, Ruse, Burgas, and Plovdiv. The production of chemicals and rubber is mainly centred around Sofia, Dimitrovgrad, Varna, Reka Devnya, and Plovdiv.

Since the 1960s three other industries have had marked regional development: food, beverage, and tobacco processing; textiles; and tourism. While food processing and beverage production are found throughout the country, three main industrial regions may be defined. The first, in the south, includes the towns of Plovdiv, Krichim, Pazardzhik, Asenovgrad, and Pŭrvomay, which primarily specialize in canning and tobacco processing. The second region, in northern Bulgaria (Gorna Oryakhovitsa, Veliko Tŭrnovo, and Lyaskovets), concentrates on canning, sugar refining, and meat processing. A third region to the northwest (Pleven, Dolna Mitropoliya, and Cherven Bryag) has become important for flour, paste products, poultry processing, canning, sugar refining, and the processing of vegetable oils. Fishing and fish breeding have also become important industries.

Textile industries before World War II were mainly to be found where the demand for textiles was constant (Sofia, Plovdiv, Varna) or where raw materials were available (Sliven, Vrasta). The industries used locally produced wool, cotton, or silk. Under the republic's five-year plans, large new mills were built at Sofia, Sliven, and Plovdiv, and the total output of textile fabrics rose tremendously.

Tourism in Bulgaria has grown markedly since the 1960s. In addition to the popular Black Sea resorts, tourists are attracted to historical centres such as Sofia, Plovdiv, and Rila Monastery and to winter sports centres such as Borovets in the Rhodope Mountains.

John Launois—Black Star

Harvesting wheat on a state farm near Tolbukhin, in the Dobruja region, northeastern Bulgaria.

The major source of energy in Bulgaria is the Maritsa lignite field, which provides fuel to large thermoelectric plants at Dimitrovgrad and Maritsa-Iztok; there are thermal power stations also at Pernik, Sofia, Plovdiv, and Burgas. The country's first nuclear power station, at Kozloduy, was constructed with Soviet aid. It began operation in 1974 and has since been expanded. Transmission lines carry electric power to most parts of the country.

Agriculture. A cooperative movement developed in the agricultural sector in Bulgaria before World War II, and after the war it spread further, until most of the arable land was in cooperative farms. The cooperative and state farms then merged into large state and collective units. Agriculture grew unevenly, however, and in 1970–71 these units were consolidated into the larger groupings called agro-industrial complexes to take better advantage of integrated systems of automation, supply, and marketing. Agriculture accounts for less than one-fifth of the national income in Bulgaria.

Crops

Cereal crops are grown on almost three-fifths of the sown land. Wheat is by far the most important, followed by corn (maize) and barley; rye, oats, soybeans, and rice are also grown. Industrial crops are important, especially tobacco, which is a good-quality Oriental type and is grown mainly in the south. Sunflower seed is the chief oil-seed crop; after extraction of the oil, the pulp is made into cattle feed. Sunflowers, like sugar beets, grow mainly in the north. Bulgaria has become a leading exporter of grapes and tomatoes. Stock breeding has increased since World War II, and productivity has improved; cattle, sheep, pigs, and poultry are raised.

Finance. As of 1990, the Bulgarian National Bank, the principal banking institution, was to undergo significant reforms. Its 130 branches were to be reorganized to have more operational independence. Sixty separate commercial banks (about half with their own subsidiaries) were to be established for better links between banks and state enterprises. Shares in local enterprises were to be offered for sale in some commercial banks.

Trade. Foreign trade is a state monopoly. Almost two-thirds of all exports are capital goods, such as machinery and equipment, and one-quarter are consumer goods, mainly of agricultural origin (such as fresh and canned fruits, wine, honey, cigarettes, dairy products, and meat and live animals). About two-fifths of all imports are capital goods. There has been some decline in Bulgaria's imports from the Council for Mutual Economic Assistance member countries, while exports to these states have remained steady, creating a trade surplus. Exports to the West have risen, and imports remained steady.

Transportation. *Roads and railways.* The development of the Bulgarian economy has required the expansion of the transportation system. Road transport accounts for a large percentage of all freight carried (on a tonnage basis) as well as most passenger traffic. Roads and railways together account for all but a small fraction of both freight and passengers carried. The E-5 European International Highway, crossing the country toward Istanbul, has undergone reconstruction. The main railway lines connect Sofia with the Black Sea coast.

Airports

Water and air transport. The Danube is used for both internal and international traffic, Ruse, Svishtov, and Lom being the main river ports. The chief seaports are Varna and Burgas, and there is a regular international merchant service on the Black Sea. Bulgaria has international airports at Sofia, Varna, and Burgas, and internal air services have grown; scheduled international service has increased considerably.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Following the overthrow of the veteran Communist Party leader Todor Zhivkov in November 1989, the party gave up its guaranteed right to rule, adopted a new manifesto, streamlined its leadership, and changed its name to the Bulgarian Socialist Party. Attempts to create a broadly based national unity government prior to special elections in mid-1990 failed because the Agrarian

Party (formerly the sleeping coalition partner of the Communist Party) and the multiparty Union of Democratic Forces refused to join the Socialists. Bulgaria's first free elections in 40 years, held on June 10 and 17, 1990, defied the predictions of most observers by granting the Socialists a comfortable majority in the National Assembly, which was to draft a new constitution.

The highest constantly functioning organ of state power is the State Council (Ministerski Sŭvet), which is responsible to the National Assembly. It consists of a president, a varying number of vice presidents, a secretary, and a varying number of members, all elected from the membership of the National Assembly. The State Council manages the day-to-day operations of the organizational and political functions of the country based on decisions made by the National Assembly.

The Council of Ministers is the official executive and administrative body and as such comprises the government of Bulgaria. It consists of a chairman (the country's prime minister), deputy chairman, ministers, and heads of departments (who may be given the rank of minister by the National Assembly). The council is elected from the membership of the National Assembly and is responsible to the Assembly.

The People's Councils express state power at the local government level. The members of the People's Councils are elected for two and a half years. They are responsible for the economy, health, social welfare, communal activities, and cultural and educational facilities within the bounds of their allotted territory.

Political parties

Elections to the various representative bodies are open to all citizens who have attained the age of 18. Prior to the reforms of 1990, the ruling party was the Bulgarian Communist Party (Bŭlgarska Komunisticheska Partiya, or BKP), founded in 1891 as the Bulgarian Social Democratic Party. The other party that was permitted to function was the Bulgarian Agrarian People's Union (Bŭlgarski Zemedelski Naroden Sŭyuz, or BZNS), founded as a peasant organization in 1899. Under the Union of Democratic Forces, founded in 1990, labour, religious, environmental, and other goups joined opposition parties to challenge the Communists, who now called themselves Socialists.

Justice. The judiciary is part of the government structure, ranking below the executive and legislative branches. The system of court organization distinguishes between general (district) courts, special courts (for the armed forces and transport facilities), and a Supreme Court. Supreme Court judges are elected by the National Assembly for terms of five years, judges of district courts for three years; judges of people's courts, on the lowest level, are elected by the general electorate for the same three-year period. All judges are subject to removal by the agency that elected them. The chief public prosecutor is elected for a five-year term by the National Assembly and is empowered to appoint and dismiss deputies and other public prosecutors in the country.

Armed forces and law enforcement. The general management of the Bulgarian military forces is the responsibility of the State Council; the Bulgarian People's Army, consisting of the Ground Forces, the Air Forces, and the Navy, is under the immediate command of the Ministry of National Defense. The term of obligatory national service is two years, except in the naval forces, in which it is three. The army is integrated with the forces of other Warsaw Treaty Organization member countries. The armed forces also include the Border Guards, the Labour Corps, and the Ministry of Transport Corps. The People's Militia and the Organization for Assistance in Defense are paramilitary organizations that also serve as local police. The state security police deal with crimes against national security or those presenting a threat to society.

Education. Primary and secondary education is the responsibility of a hierarchy of educational councils. Higher education is governed by the Ministry of Public Education. Education is free at all levels, and an eight-year elementary program is obligatory for children. Since 1959, general education has included polytechnical subjects and vocational training. The "Kliment Ohridski" Sofia University (founded in 1888 as the Sofia Higher Institute) is the oldest body of higher learning in Bulgaria and was the only university until 1971, when teacher-training institutes in Plovdiv and Veliko Tŭrnovo were elevated to university status. There are numerous technical institutes as well as schools for the arts, including an institute of music, a conservatory, an institute of fine arts, and an institute of dramatic art and cinematography.

Health and welfare. The health service, directed by a government ministry, provides free medical assistance of all types. A number of specialized establishments treat tuberculosis patients, and a highly developed network of pediatric facilities cares for children. Medical social services are supported by an insurance plan that consists of a consolidated fund of contributions from state and cooperative enterprises and institutions and similar bodies. It is directed by the Confederation of Independent Bulgarian Trade Unions and covers practically all of the working population.

Urban housing shortage

Housing. A sizable fraction of the total capital investment in Bulgaria goes into housing, reflecting a program adopted in 1970 to try to alleviate the severe housing shortage caused by rapid urbanization. Another portion is spent on individual housing construction, with the state granting to individuals loans that can be repaid over periods of up to 25 years. More than 80 percent of the housing is owner-occupied.

Social conditions. Because wages and many retail prices are controlled by the government, they are kept relatively stable in relation to one another. The highest annual wages are those in the building trades, the lowest in agriculture. Monthly children's allowances, pensions, and sick pay in some cases add considerably to the income from wages alone.

The rapid industrialization since World War II has had noticeable socioeconomic effects: more than one-third of the people are now engaged in industry, while only one-fifth are in agriculture. The number of people employed in education, in culture and art, and in the health services has risen to more than one-seventh of the active work force. The percentage of women workers has risen to almost half of the total labour force, and women are especially concentrated in finance, banking, and insurance.

CULTURAL LIFE

Contemporary Bulgarian culture is a lively blend of the millennium-old folk traditions of the region; a more formal culture, which played a vital role in the emergence of national consciousness under Ottoman rule and in the development of a modern state; and, since World War II, the development of a strong Socialist element.

Because Bulgaria's population is largely homogeneous, the degree of cultural variation even at the regional level is small. The state encourages cultural development at all levels of society and supports the dissemination of culture, particularly through schools, libraries, museums, publishing and the press, radio, television, and cinema.

Daily life. Outwardly, daily life in Bulgaria has been dominated since World War II by the Socialist political system based on the teachings of Karl Marx. The state has sought to inculcate a new mode of thinking and manner of action, based above all on the need for and benefit of social labour. More inwardly, however, daily life continues to be dominated by a much older tradition and cultural legacy. For example, the Bulgarian family has kept many of its traditional forms of organization. Many households consist of an extended family comprising parents and one of their married sons—usually the youngest—or daughters.

Religious functions are entirely separate from state functions, the postwar constitution prohibiting the use of religion or religious organizations for political purposes. On the other hand, the Bulgarian state tolerates religion and does not attempt to take away the political rights of religious believers; in normal circumstances there is no attempt to persecute religious believers. Nevertheless, the state education system encourages a scientific and atheistic education.

The arts. *Visual arts.* The early impetus of Bulgarian traditions in the arts was cut short by the Ottoman occupation in the 14th century, and many of the early masterpieces were destroyed. Artistic life emerged again in Bulgaria during the national revival in the 18th century: among the most influential works were the secular and realist paintings of Zahari Zograph (1810–53) and Hristo Tsokev (1847–83). At the end of the 19th century, Bulgarian painters such as Anton Mitov (1862–1930) and the Czech-born Ivan Mrkvička (1856–1938) produced memorable works, many of them depicting the daily life of the Bulgarian people. The early part of the 20th century saw further development of both style and subject matter, and the foundations were laid for later artists such as Vladimir Dimitrov (1882–1960), an extremely gifted painter specializing in the rural scenes of his native country; Tsanko Lavrenov (1896–1978), a noted graphic artist and art critic who also painted scenes of old Bulgarian towns; Zlatyo Boyadjiev (1903–76), noted for his village portraits; and Ilya Petrov (1903–75), who painted scenes and themes from Bulgarian history. After World War II, Socialist Realism began to dominate Bulgarian artistic circles; its influence is seen in the broad historical themes adopted by modern artists in genres ranging from cartoons to still-life paintings and regional landscapes.

Bulgarian classical music

Music. The first performances of Bulgarian classical music date from the 1890s, and the earliest Bulgarian opera, by Emanuil Manolov (1860–1902), was performed in 1900. He, along with other Bulgarian composers, concentrated on solo and choral vocal works, although between World War I and World War II several symphonies and works for ballet, in addition to choral and opera works, were created by such composers as Lyubomir Pipkov (1904–74), Petko Stainov (1896–1977), and Pancho Vladigerov (1899–1978). The modern Bulgarian composers have experimented with new tonality in vocal and instrumental music. Recordings and concert tours abroad have won new audiences for traditional Bulgarian vocal music.

Theatre. The Bulgarian theatre remains a popular source of entertainment. In many cases political themes are adopted for modern plays, as evidenced by the works of Georgi Karaslavov (1904–80) in his depictions of the Bulgarian countryside and the socialist struggle for social justice. World classics and modern foreign dramas are also produced.

Literature. Bulgaria's literary tradition may be traced to the 9th century, when saints Cyril and Methodius created the alphabet for Old Bulgarian (Old Church Slavonic). During Ottoman domination, literature could only be preserved in the monasteries and churches. With the national revival in the 18th century, Bulgarian literature flowered once more and was used as a subtle spiritual means of fighting for the country's liberation.

Motion pictures. Bulgaria's motion-picture industry has expanded considerably since World War II. Bulgarian audiences take interest in both domestic and foreign films, and the country's feature and documentary films have a wide export market. Animated cartoons also are popular. Many of the motion pictures are produced at the Cinema Centre near Sofia.

Cultural institutions. Bulgaria has almost 10,000 libraries, with a total holding of about 110,300,000 volumes, and more than 200 museums. There are also a few thousand *chitalishte*s, cultural centres similar to reading rooms, found in even the smallest villages. The major museums in Sofia are the National Archaeological Museum, National Ethnographical Museum, National Museum of the Revolutionary Movement of Bulgaria, and the National Museum of Bulgarian-Soviet Friendship. Plovdiv also has some well-known museums, particularly for archaeology and ethnography, and other important archaeological museums are in Stara Zagora, Burgas, and Varna. Regional specialist museums are also popular. The major state libraries include the Cyril and Methodius National Library and the Central Library of the Bulgarian Academy of Sciences, both in Sofia.

The highest research institution in the country is the Bulgarian Academy of Sciences. It is involved in international cooperative projects and has cultural treaties with other eastern European countries and with some research institutes in the West.

Recreation. *Festivals.* Bulgaria observes many festivals and competitions. Examples are the International Folklore Festival, held early in August in Burgas, and the Varna Summer International Festival, primarily a music festival, held in July. Historical plays are popular, particularly when staged outdoors in summer against the natural backdrop of important monuments or buildings associated with events in the country's history. The local festivals provide an opportunity for new musical and literary works to be performed.

The blossoming of the roses in the Karlovo and Kazanlŭk valleys is celebrated in May and June; the oil-bearing roses are collected for production of attar of roses, which is exported worldwide.

Sport and physical education. The state-run Union of Physical Education and Sport is responsible for building and maintaining sports facilities, training coaches, and overseeing athletics at all educational levels. Athletes have considerable sporting facilities at their disposal, including Vasil Levski Stadium, Universiada Hall, Festivalna Hall, and Students' Sports Complex in Sofia and the Palace of Sport and Culture in Varna; many regional centres also have sports complexes for local use. The major sporting event in the country is the National Spartakiad, which organizes mass participation of sports teams, clubs, and individuals in a variety of sporting activities.

Press and broadcasting. Broadcasting is the responsibility of the Committee for Television and Radio. In addition to national and regional programs Bulgarian Radio also broadcasts in several languages to foreign countries. Bulgarian National Television produces several types of programs, including news coverage and documentaries, arts and education programs, sports broadcasts, programs for children and youth, and programs for visiting tourists. A Soviet channel also broadcasts. All mass media, including printed matter, are strictly censored. (F.W.C.)

For statistical data on the land and people of Bulgaria, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

History

Evidence of human habitation in the Bulgarian lands dates from the Middle Paleolithic period (100,000 to 40,000 BC). Agricultural communities appeared in the Neolithic period, and in the Bronze Age the lands were inhabited by Thracian tribes. The Thracians were eventually expelled or absorbed by Greek, Persian, and Roman colonies, but traces of their culture remain in their monuments devoted to horse worship and in the mummer (Bulgarian: *kuker*) tradition that still survives in southwestern Bulgaria. In Roman times Bulgaria was divided between the provinces of Moesia (to the north of the Balkan Mountains) and Thrace (to the south of the Balkans) and was crossed by the main land route from the west to the Middle East. The ruins of Roman towns and settlements are numerous, and extensive sites have been excavated at Plovdiv in the southwest, Varna in the northeast, and other locations.

Traces of early Thracian habitation

THE BEGINNINGS OF MODERN BULGARIA

The Slavic invasions. The story of the modern Bulgarian people begins with the Slavic invasions of the Balkan Peninsula in the 6th and 7th centuries AD, a time when Byzantium was absorbed in prolonged conflict with Persia and could not resist the incursions from the north. Ancient sources refer to two Slavic tribes north of the Danube at this time, the Slavenae and the Antae. Evidence suggests that the Slavenae, to the west, were the ancestors of the Serbs and Croatians, while the Antae moved into the regions of Bulgaria, Macedonia, and northern Greece. The Slavic tribes tilled the soil or practiced a pastoral way of life and were organized in patriarchal communities.

Arrival of the Bulgars. The name Bulgaria comes from the Bulgars, a Turkic people who are first mentioned in the sources toward the end of the 5th century AD. Living at that time in the steppes to the north of the Black Sea, the Bulgar tribes were composed of skilled, warlike horsemen governed by khans (chiefs) and boyars (nobles). The Bulgars were subdued by the Avars in the 6th century, but in 635 Khan Kubrat led a successful revolt and organized an independent tribal confederation. After Kubrat's death in 642 the Bulgars were attacked by the Khazars and dispersed. According to Byzantine sources, the Bulgars split into five groups, each under one of Kubrat's sons. One of these, Asparukh (or Isperikh), moved into Bessarabia (between the Dnestr and Prut rivers) and then crossed to the south of the Danube, where his people conquered or expelled the Slavic tribes living north of the Balkan Mountains. The Byzantine emperor Constantine IV led an army against the Bulgars but was defeated, and in 681 Byzantium recognized by treaty Bulgar control of the region between the Balkans and the Danube. This is considered to be the starting point of the Bulgarian state.

THE FIRST BULGARIAN EMPIRE

Asparukh and his successors established their court at Pliska, northeast of modern Shumen, and a religious centre at nearby Madara. Archaeological evidence suggests that the Bulgars kept their settlements distinct from those of the Slavs, from whom they accepted tribute. They maintained a mixed pastoral and agricultural economy, although much of their wealth continued to be acquired through warfare. Asparukh's successor, Tervel (701–718), helped to restore the emperor Justinian II to the Byzantine throne and was rewarded with the title "caesar." On the whole, however, relations with Byzantium were hostile, and the 8th century was marked by a long series of raids and larger campaigns in which the Byzantine forces were usually victorious. Bulgaria recovered under Khan Krum (803–814), who, after annihilating an imperial army, took the skull of the emperor Nicephorus I, lined it with silver, and made it into a drinking cup. Under Krum's successors Bulgaria enjoyed an extended period of peace with Byzantium and expanded its control over Macedonia and what is now western Yugoslavia.

Assimilation of Bulgars by Slavs. Internally, the 8th and 9th centuries saw the gradual assimilation of the Bulgars by the Slavic majority. There are almost no sources that describe this process, but it was certainly facilitated by the spread of Christianity, which provided a new basis for a common culture. Boris I of Bulgaria (852–889) was baptized a Christian in 864, at a time when the conflict between the Roman Church and the Eastern Church in Constantinople was becoming more open and intense. Although Boris' baptism was into the Eastern Church, he subsequently wavered between Rome and Constantinople until the latter was persuaded to grant de facto autonomy to Bulgaria in church affairs. The spread of Christianity was facilitated by the work of saints Cyril and Methodius, who had invented an alphabet in which to write the Slavic language (known as Old Church Slavonic or Old Bulgarian) and who had developed a Slavonic liturgy in Moravia. When Moravia committed to Rome and expelled the disciples of Cyril and Methodius, many of them resettled in Bulgaria, where they were welcomed by Boris and undertook the translation of church books and the training of priests. St. Clement and St. Naum are credited with preparing more than 3,000 priests, and they established an important church and educational centre on the shores of Lake Ohrid (Okhrid) in Macedonia.

Spread of Christianity

Bulgaria's conversion had a political dimension, for it contributed both to the growth of central authority and to the melding of Bulgars and Slavs into a unified Bulgarian people. Boris adopted Byzantine political conceptions, referring to himself as ruler "by the grace of God," and the new religion provided justification for suppressing those boyars of Bulgar origin who clung to paganism and the political and social order with which it was linked. In 889 Boris, whose faith apparently was deep and genuine, abdicated to enter a monastery. When his eldest son, Vladimir, fell under the influence of the old boyars and attempted to reestablish paganism, Boris led a coup that overthrew him. After Vladimir was deposed and blinded, Boris convened a council that confirmed Christianity as the religion of the state and moved the administrative capital from Pliska to the Slavic town of Preslav. The council conferred the throne on Boris' third son, Simeon, and Boris retired permanently to monastic life.

Reign of Simeon I. The reign of Simeon I (893–927) marked the high point of the first medieval Bulgarian state. Educated in Constantinople and imbued with great respect for the arts and Greek culture, Simeon encouraged the building of palaces and churches, the spread of monastic communities, and the translation of Greek books into Slavonic. Preslav was made into a magnificent capital that observers described as rivaling Constantinople. The artisans of its commercial quarter specialized in ceramics, stone, glass, wood, and metals, and Bulgarian tile work in the "Preslav style" surpassed its contemporary rivals and was eagerly imported by Byzantium and Kievan Rus.

Simeon was also a gifted military leader. His campaigns extended Bulgaria's borders, but he ultimately dissipated the country's strength in an effort to take Constantinople. When he died he was master of the northern Balkans, including the Serbian lands, and styled himself "Tsar of the Bulgars and Autocrat of the Greeks," but his country was near exhaustion.

Under Simeon's successors Bulgaria was beset by internal dissension provoked by the spread of Bogomilism (a dualist religious sect) and by assaults from Magyars, Pechenegs, the Rus, and Byzantines. In the campaign of 1014 the Byzantine emperor Basil II won a decisive victory, after which he blinded as many as 15,000 prisoners taken in the battle. (For this act he became known as Basil Bulgaroctonus, or Basil, Slayer of the Bulgars.) Bulgaria lost its independence in 1018, and for more than a century and a half, until 1185, remained subject to Byzantium.

THE SECOND BULGARIAN EMPIRE

With the collapse of the first Bulgarian state, the Bulgarian Church fell under the domination of Greek ecclesiastics who took control of the see of Okhrid and attempted to replace the Bulgarian Slavic liturgy with liturgy in the Greek language. Bulgarian culture was by this time too deeply rooted to be easily removed, and the Byzantine Empire, beset by the attacks of the Seljuq Turks and the disturbances of the crusaders, lacked the power to support a more forcible Hellenization. In 1185 the brothers Ivan and Peter Asen of Tŭrnovo launched a revolt to throw off Byzantine sovereignty. The Asen brothers defeated the Byzantines and forced Constantinople to recognize Bulgarian independence. Their brother and successor, Kaloyan (1197–1207), briefly accepted the supremacy of Rome in church affairs and received a royal crown from the pope. But when the patriarch at Constantinople again recognized the independence of the Bulgarian Church, Kaloyan reverted to Orthodoxy.

The Second Bulgarian Empire, with its centre at Tŭrnovo, reached its height during the reign of Tsar Ivan Asen II (1218–41). Bulgaria was then the leading power in the Balkans, holding sway over Albania, Epirus, Macedonia, and Western Thrace. During this period the first Bulgarian coinage appeared, and in 1235 the head of the Bulgarian Church received the title of patriarch.

The successors of Ivan Asen II could not match his ability. Moreover, Bulgaria was beset by Mongol attacks from the north and by internal upheavals brought on by the growing burdens placed on the peasantry by the powerful nobles. The great peasant revolt of 1277–80 briefly allowed the swineherd Ivaylo to occupy the royal throne at Tŭrnovo before he was defeated with the aid of the Byzantines. The Asen dynasty died out in 1280 and was followed by two others, both of Cuman origin, neither of which succeeded in restoring central authority. The declining state reached its nadir in 1330 when Tsar Mikhail Shishman was defeated and slain by the Serbs at the battle of Velbuzhd (modern Kyustendil). Bulgaria lost its Macedonian lands to the Serbian Empire of Stefan Dušan, which then became the dominant Balkan power. Bulgaria appeared to be on the point of disintegration into feudal states when the invasions of the Ottoman Turks began.

OTTOMAN RULE

The Ottoman Turks first entered the Balkans as mercenaries of Byzantium in the 1340s, and they returned as invaders in their own right during the following decade. Between 1359 and 1362 Sultan Murad I wrested much of

Thrace from Byzantine control and captured Adrianople, commanding the route up the Maritsa Valley into the heart of the Bulgarian lands. In 1364 the Turks defeated a crusade sent by Pope Urban V to regain Adrianople, but not before the crusaders committed so many atrocities against the Orthodox Christians that many Bulgarians came to regard Turkish rule as preferable to alliance with the Roman Catholic West. Although Ivan Shishman, Bulgaria's last medieval tsar, declared himself a vassal of Murad in 1371, the Ottomans continued to seek complete domination. Sofia, in the west, was seized in 1382, and Shumen, in the east, fell in 1388. A year later the defeat of the Serbs on Kosovo Plain sealed the fate of the entire Balkan Peninsula. In 1393, after a three-month siege, Tŭrnovo was taken and burned. Ivan Shishman died in Turkish captivity three years later. With the capture of a rump Bulgarian kingdom centred at Bdin (Vidin) in 1396, the last remnant of Bulgarian independence disappeared.

The "Turkish yoke." The five centuries from 1396 to 1878 are engraved in Bulgarian consciousness as the era of the "Turkish yoke," traditionally seen as a period of darkness and suffering. Both national and ecclesiastical independence were lost. The Bulgarian nobility was destroyed—its members either perished, fled, or accepted Islām and Turkicization—and the peasantry was enserfed to Turkish masters. The "blood tax" took a periodic levy of male children for conversion to Islām and service in the Janissary Corps of the Ottoman army. The picture was not entirely negative, however. Once completed, the Turkish conquest included Bulgaria in a "pax ottomanica" that was a marked contrast to the preceding centuries of war and conflict. While Ottoman power was growing or at its height it provided an acceptable way of life for the Bulgarian population. It was only when the empire was in its decline and unable to control the depredations of local officials or maintain reasonable order that the Bulgarians found Ottoman rule unbearable.

Bulgaria did not change radically in its religious or ethnic composition during the Ottoman period, for the Turks did not attempt to forcibly populate Bulgaria with Turks or to convert all Bulgarians to Islām. With the exception of the people of the Rhodope Mountains who did convert (and thence were called Pomaks), the Bulgarian population remained within the Orthodox Church. Although Turkish administrators were established in the towns and countryside, Turkish peasants were not settled in Bulgaria in large numbers, and those who did immigrate were concentrated in the southern and eastern parts of the country and in some of the valleys of Macedonia and Thrace. In the 15th and 16th centuries Turkish authorities permitted the immigration of Jewish refugees from the Christian West. While the majority were resettled in Constantinople and Salonika, most Bulgarian towns acquired small Jewish communities.

Ottoman administration. At the time Bulgaria was conquered, the Ottoman Empire was divided into two parts for administrative purposes. Bulgaria was part of the European section, called Rumelia, headed by a *beylerbey* ("lord of lords") who resided in Sofia. As the empire expanded, this system proved inadequate, and in the 16th century it was replaced by territorial divisions called *vilayet*s (provinces), further subdivided into *sanjak*s (districts). The borders of these units changed many times over the centuries. Bulgarian lands were assigned as fiefs to Turkish warriors, or *spahi*s, who could impose taxes and other obligations on the subject population. Fiefs were also given to governors and other officeholders to provide their income, and lands in the form of *vakıf*s—designated for the support of religious, educational, or charitable enterprises—were assigned to specific institutions. The *spahi* had no right of lordship or justice over the peasants living in his fief, and the Bulgarians frequently retained their traditional village administration and the customs of local law with regard to issues in which Turkish interests were not involved.

Decline of the Ottoman Empire. The decline of the Ottoman Empire was marked by military defeats at the hands of Christian Europe and by a weakening of central authority. Both of these factors were significant for developments in Bulgaria. As the empire was thrown on the defensive, the Christian powers, first Austria and then Russia, saw the Bulgarian Christians as potential allies. Austrian propaganda helped to provoke an uprising at Tŭrnovo in 1598, and two others occurred in 1686 and 1688 after the Turks were forced to lift the siege of Vienna. Under Catherine the Great, Russia began to assert itself as the protector of the Orthodox population of the Ottoman Empire, a claim that the Sublime Porte (as the government of the empire was called) was forced to recognize in the Treaty of Küçük Kaynarca in 1774. Of greater significance, however, was the inability of the central government to keep the *spahi*s and local officials under control. During the 17th and 18th centuries the *spahi*s succeeded in converting their fiefs to *chifliks*, hereditary estates that could not be regulated by the government. Owners of *chiflik*s were free to impose higher obligations on the peasantry or to drive them off the land. Turkish refugees from lands liberated by Christian states were frequently resettled on *chifliks* in Bulgaria, increasing the pressure on the land and the burden on the peasantry. Occasionally, Turkish refugees formed marauding bands that could not be subdued by central authority and that exacted a heavy toll from their Christian victims. One response among the Bulgarians was a strengthening of the *haiduk* tradition. The *haiduk*s were guerrillas—some would say bandits—who took to the mountains to live by robbing the Turks. Although the *haiduk*s lacked a strong sense of national consciousness, they kept alive a spirit of resistance and gave rise to legends that inspired later revolts.

THE NATIONAL REVIVAL

In the 19th century growing Bulgarian discontent was given direction in a movement of national revival that restored Bulgarian national consciousness and prepared the way for independence. The social foundation of this movement was produced by the quickening of economic life in the late 18th and early 19th centuries and by the influence of the Enlightenment and the French Revolution, echoes of which, however faint, were heard among the people. A growing demand for cotton cloth and other products stimulated urban development. A number of Bulgarian merchant houses were founded, and artisans in the towns began to form guild organizations (*esnafi*). The latter played an important role in sponsoring schools and providing scholarships for young Bulgarians to study abroad.

Paisiy of Khilendar

The monk Paisiy of the Khilendar Monastery on Mount Athos is recognized as the founder of the national revival. Little is known of his life except that he came from a merchant family in Bansko, a town in southwestern Bulgaria that maintained commercial relations with Vienna. In the 1760s Paisiy used texts preserved on Mount Athos to write his "Slaveno-Bulgarian History." It reminded Bulgarians of the greatness of their past empires and called on them to foreswear foreign tongues and customs and to take pride in their race. Sofroniy, Bishop of Vratsa, helped to spread Paisiy's influence. In his own writings he stressed the importance of education, without which his people would remain, in his words, "dumb animals."

Spread of education. The spread of education was in fact the centrepiece of the Bulgarian national revival. In 1835 Vasil Aprilov founded a Lancasterian school, based on the monitorial system of instruction, in Gabrovo. With the monk Neofit Rilski (Neophyte of Rila) as its teacher, it was the first school to teach in Bulgarian. Its work was facilitated by the appearance of a Bulgarian publishing industry and a small but influential periodical press. By the 1870s the guilds, town and village councils, and wealthy groups and individuals founded some 2,000 schools in Bulgaria, each providing free education. The school was supplemented with the *chitalishte,* or "reading room," an institution that first appeared in Svishtov in 1856 but soon spread throughout the country. More than just a small library, the *chitalishte* staged lectures, meetings, plays, concerts, debates, and social events. It was of immense importance for those who did not acquire formal education.

Cultural movement against Greek influence. The cultivation of Bulgarian national consciousness was initially a cultural rather than a political movement. Consequently, it was directed more against the "cultural yoke" of the Greeks than the "political yoke" of the Ottoman Empire. After the Turkish conquest of the Balkans, the Greek patriarch had become the representative of the *Rūm millet,* or the "Roman nation," which comprised all the subject Christian nationalities.

The desire to restore an independent Bulgarian church was a principal goal of the national "awakeners." Their efforts were rewarded in 1870 when the Sublime Porte issued a decree establishing an autocephalous Bulgarian church, headed by an exarch, with jurisdiction over the 15 dioceses of Bulgaria and Macedonia. Although the Greek patriarch refused to recognize this church and excommunicated its adherents, it became a leading force in Bulgarian life, representing Bulgarian interests at the Sublime Porte and sponsoring the further expansion of Bulgarian churches and schools. After the liberation of 1878 it provided a powerful means of spreading Bulgarian national feeling in Macedonia.

The national revolution. The creation of the Bulgarian exarchate was the high point of the national revival as a cultural movement. The inability of the Sublime Porte to maintain order or to carry through its program of reform known as Tanzimat (1839–76), combined with the examples of Greek and Serbian independence, engendered an explicitly revolutionary movement among the Bulgarians. Inspired by the *haiduk* tradition, Georgi Rakovski formed a Bulgarian legion on Serbian territory in 1862 to send armed bands to harass the Turks in Bulgaria. In 1866 Lyuben Karavelov and Vasil Levski created a Bulgarian Secret Central Committee in Bucharest to prepare for a national uprising. It dispatched "apostles" into Bulgaria to spread the message among the people. Levski was captured during one such mission and was hanged in Sofia. He is considered to be the greatest hero of the revolutionary movement.

Uprising of 1876

Against the background of a wider Balkan crisis, the Bulgarian revolutionary committees laid plans for a nationwide uprising in 1876. The April Uprising began prematurely and was violently put down. The atrocities committed against the civilian population by irregular Turkish forces, including the massacre of 15,000 Bulgarians near Plovdiv, increased the Bulgarian desire for independence and outraged public opinion in Europe. A conference of European statesmen proposed a series of reforms, but when the Sultan refused to implement them Russia declared war. In the ensuing campaign Bulgarian volunteer forces fought alongside the Russian army, earning particular distinction in the epic battle for the Shipka Pass.

Treaty of San Stefano (1878). Advancing to the outskirts of Constantinople, the Russians dictated the Treaty of San Stefano, which called for a large independent Bulgaria stretching from the Danube to the Aegean and from the Vardar and Morava valleys to the Black Sea. The boundaries stated in the treaty, signed on March 3, 1878, represented the fulfillment of Bulgaria's territorial aspirations and remained for generations the national ideal of the people. But the creation of a large Bulgaria, perceived as an outpost of Russian influence in the Balkans, was intolerable to Austria-Hungary and Britain, and they forced a revision of the Treaty of San Stefano a few months later at the Congress of Berlin. The new Treaty of Berlin (July 13, 1878) created a much smaller Bulgarian principality, autonomous but under the sovereignty of the Sublime Porte, in the territory between the Danube and the Balkan Mountains. To the south, the treaty created the autonomous province of Eastern Rumelia, subject to the sultan but with a Christian governor. Macedonia was returned entirely to the Ottoman Empire. The treaty also stipulated that Bulgaria would elect an assembly of notables to meet at Tŭrnovo to prepare a constitution and to choose a prince who would be confirmed by the powers.

Land reform after liberation

The liberation of Bulgaria from Turkish rule also functioned as a land reform, for Russian occupation authorities and subsequent Bulgarian governments confiscated the Turkish estates and sold them in small parcels to the peasantry. Bulgaria began its independence as a nation of smallholders with one of the most equalitarian land distributions in Europe.

THE PRINCIPALITY

By the time the constituent assembly convened in Tŭrnovo in February 1879, conservative and liberal political tendencies had emerged and rapidly coalesced into parties. The Liberal Party, under Dragan Tsankov, Karavelov, and Petko Slaveikov, dominated the assembly and created a constitution that was one of the most democratic in Europe. It provided for a single National Assembly elected by universal male suffrage, guarantees of civil rights, and strict limits on the power of the prince.

Political divisions under Alexander of Battenberg. The democratic character of the constitution and sympathies of Liberal Party political leaders were at variance with the views of Bulgaria's first prince, Alexander of Battenberg (a nephew of the Tsar), and with those of the Russian advisers who played a large role in his court. The Prince first formed a Conservative ministry, but he was forced by popular agitation to form a Liberal government under Tsankov. Tsankov's government undertook the construction of judicial and state apparatuses and put an end to the depredations of brigands who had remained active in the mountains after the war. But when the Liberals showed, in the eyes of the Prince, insufficient respect for the institution of monarchy and engaged in a campaign to diminish Russian influence, Alexander dismissed them in favour of a Conservative government led by General Casimir Erenroth, a Finn in Russian service, who had earlier been charged with setting up the Bulgarian Army. Erenroth used rigged elections to select a Grand National Assembly, which agreed in 1881 to suspend the constitution and invest the Prince with absolute powers for seven years. A period of dictatorship followed under the Conservatives and the Russian generals Leonid N. Sobolev and Alexander V. Kaulbars. Alexander, however, soon found his Russian allies harder to deal with than their Liberal predecessors. In September 1883 he compromised with his opponents, dismissed the Russians, restored the constitution, and accepted a coalition Conservative–Liberal government, but the coalition was soon supplanted by an entirely Liberal government under Petko Karavelov, the brother of Lyuben.

Eastern Rumelia's union with Bulgaria

Meanwhile, popular sentiment for union with Bulgaria had been growing in Eastern Rumelia, and the restoration of the constitution provided the Eastern Rumelian allies of Bulgaria's Liberal Party with the stimulus to prepare for a seizure of power in Plovdiv. In September 1885, with the prior approval of Prince Alexander, they staged a bloodless coup d'état and declared the unification of the two states. Turkey did not resist, but Russia, incensed by such independence of action in its diplomatic sphere, refused to approve, and Tsar Alexander III ordered the withdrawal of all Russian officers and advisers in the Bulgarian Army. In these circumstances, King Milan of Serbia, stating that the balance of power in the Balkans was endangered by Bulgarian unification, suddenly declared war. The Serbs advanced as far as Slivnitsa, where they were met and defeated by the untrained Bulgarian army under Prince Alexander's command. Bulgarian forces pursued the Serbs across the frontier but were stopped by the threat of Austrian intervention. Peace and the status quo were restored by the Treaty of Bucharest (March 3, 1886) and the convention of Tophane (April 5, 1886). Prince Alexander was appointed governor general of Eastern Rumelia and the Eastern Rumelian administrative and military forces were merged with those of Bulgaria.

Prince Alexander had little time to enjoy the fruits of his popular triumph. On Aug. 21, 1886, a group of Russophile conspirators and military officers whom Alexander had passed over for promotion seized the Prince in his palace, forced him to sign a statement of abdication, transported him out of the country, and handed him over to the Russians at the Danube port of Reni. The conspiracy was countered, however, by Stefan Nikolov Stambolov, president of the National Assembly, and by Lieutenant Colonel Sava Mutkurov, commander of the Plovdiv garrison, who took control of Sofia and recalled the Prince. Alexander was not detained by the Russians, but he refused to remain in Bulgaria without Russian approval. When the Tsar refused to give it, Alexander abdicated on September 7, appointing a regency composed of Stambolov, Mutkurov, and Petko Karavelov. The regency was successful in preserving order but had great difficulty finding a new prince, for few wished to assume the throne in the face of Russian hostility. A willing candidate was at last found in the person of 26-year-old Prince Ferdinand of Saxe-Coburg-Gotha, a grandson of Louis-Philippe of France, who was then serving as an officer in the Austrian army. Ferdinand was elected prince by the Grand National Assembly in July 1887.

Prince Ferdinand's rule. Because Russia declared Ferdinand a usurper, Europe withheld recognition, the bishops of the Holy Synod would not pay him homage, and conspiracies flourished. However, Stambolov, as prime minister from 1887 to 1894, ruthlessly suppressed all opposition. Stambolov, who was recognized as one of Europe's "strongmen," stabilized Bulgaria's international position, but his methods, which amounted to a virtual dictatorship, alienated much of the population. In 1894 Ferdinand unexpectedly made use of his constitutional right to dismiss Stambolov and replaced him with a Russophile government headed by a Conservative, Konstantin Stoilov. A year later the former prime minister was gunned down on the streets of Sofia, apparently by some of those he had earlier persecuted.

Dismissal and death of Stambolov

The change of course in Sofia and the death of Tsar Alexander III facilitated a reconciliation between Bulgaria and Russia. Ferdinand gained international recognition as prince, and in 1896 Tsar Nicholas II stood godfather to Ferdinand's first son when he was baptized into the Orthodox faith.

The failure of Bulgarian democracy. *Breakdown of two-party system.* From the time of the liberation, political power was exercised by a politically active elite composed of the country's few large landowners, merchants, lawyers, clergy, officers, teachers, and those educated in Bulgarian communities abroad. After the liberation these men staffed the country's civilian and military bureaucracies. By the mid-1880s their number expanded far beyond the capacity of the country to absorb them effectively, and the educational system continued to generate a large surplus of lawyers, journalists, and literati who had little hope for careers outside the government. The result was the breakdown of the two-party system, an intensified struggle for the spoils of office, a proliferation of bureaucracy, and pervasive corruption. Between 1884 and the end of the century, splits in the ranks of Conservatives and Liberals gave rise to nine parties. There were few, if any, differences in principle among them, and they were, in the words of one participant, "partnerships for the exploitation of power." The peasantry, which made up four-fifths of the population, lacked the education, organization, and experience to act as a brake on political degeneration.

Economic stagnation. The loss of free access to Ottoman markets and the lack of available capital largely offset the economic advantages gained by the liberation. Moreover, Bulgarian governments did little to encourage industrialization, preferring to use the state budget to finance the growth of bureaucracy and to maintain as powerful an army as possible in the hope of territorial expansion. Bulgarian military expenditures on a per capita basis were the highest in Europe and brought the country the reputation of being "the Prussia of the Balkans." In the countryside the long-term factors of low agricultural productivity, declining world grain prices, and steady population growth were reinforced by government policies that imposed increasing burdens of taxation and indebtedness.

The Socialist movement and the Bulgarian Agrarian Union. In the last decade of the 19th century these conditions gave rise to two movements of political protest that would significantly affect Bulgaria's future. One was the appearance of a Socialist movement that drew its inspiration from Karl Marx and the European Socialist parties. The schoolteacher Dimitŭr Blagoev, who as a student in Russia had formed the first Marxist group in that country, inspired the creation of a Bulgarian Socialist party in 1891. Given Bulgaria's minuscule proletariat, Marxist socialism had little chance to become a mass movement, but it remained an influential force among the country's intelligentsia. Of greater immediate importance was the emergence in the countryside of an alliance between the peasantry and the rural intelligentsia. This acquired organizational form in 1899 with the founding of the Bulgarian Agrarian Union (later the Bulgarian Agrarian National Union). Created as a nonpartisan interest group to promote peasant welfare, the Agrarian Union was pushed into politics by the unrest and violence that followed the government's attempt to reintroduce the Turkish tithe on agricultural production in 1900. Under the leadership of Aleksandŭr Stamboliyski, who developed a distinct agrarian ideology and program, the Agrarian Union grew rapidly and was the largest opposition party in the country on the eve of the Balkan Wars.

Leadership of Aleksandŭr Stamboliyski

Ferdinand's "personal regime." The period from Stefan Stambolov's fall in 1894 to World War I is known as the era of Ferdinand's "personal regime." By encouraging the fragmentation of the political parties and by skillfully using his powers of patronage to manipulate the party chiefs, Ferdinand became the dominant political figure in the country. In 1908, in conjunction with the Austrian annexation of Bosnia and Hercegovina, he proclaimed the de jure independence of Bulgaria from the Sublime Porte and assumed the title "tsar." Three years later a Grand National Assembly amended the constitution to give him this title officially and to grant him the right to conclude treaties with foreign states without the knowledge of the National Assembly.

Macedonia constituted the principal objective of Ferdinand's diplomacy. On July 20, 1903, the Macedonian Bulgarians revolted against Ottoman rule. Their revolt—known as the Ilinden (St. Elijah's Day) Uprising—and the brutality with which it was put down again focused attention on the problems of Turkish misrule in Macedonia, and in 1908 the revolution of the Young Turks led Balkan statesmen to believe that the time was fast approaching when Macedonia could be wrested from the empire. Greece and Serbia, however, laid claim to portions of Macedonia that Bulgarians regarded as rightfully theirs. It was the great mistake of Bulgarian diplomacy to organize a war against the Ottoman Empire without first clearly resolving these competing claims.

Unresolved claims to Macedonia

In March 1911, against the background of increasing unrest in Macedonia, Ferdinand appointed a new government under Ivan Geshov to begin negotiations for an anti-Turkish alliance. In May 1912 Bulgaria signed a treaty with Serbia providing for military cooperation but leaving a large section of Macedonia as a contested zone, the fate of which would be determined after the war. A quickly made agreement with Greece also made no provision for the future distribution of territory. An arrangement between Greece and Serbia and verbal agreements with Montenegro completed the formation of the Balkan League. Montenegro declared war on the Ottoman Empire on October 8, and the other Balkan states soon entered the conflict.

Bulgaria's participation in the First and Second Balkan wars. The successes of the Balkan League exceeded expectations. Bulgarian forces won major victories at Lozengrad (now Kırklareli) and Lüleburgaz and laid siege to Adrianople and the Çatalca line of fortifications defending Constantinople, while the Greeks took Salonika and Serbian troops won a series of battles in Macedonia. Turkey asked for an armistice, but Ferdinand insisted that the army attempt to capture Constantinople. When the assault on the Çatalca line failed, leaving the Bulgarian army in a weakened state, the Tsar agreed to the armistice, and peace negotiations began in London. On May 30, 1913, Turkey signed the Treaty of London, conceding all but a small strip of its European territory. But it proved impossible to divide the territory peacefully among the victors. Serbia and Greece insisted on retaining most of the Macedonian territory they had occupied, and Romania demanded compensation for its neutrality. When Geshov was not able to negotiate a compromise he resigned in favour of Stoyan Danev, who reflected the Tsar's desire for a military solution. On the night of June 29/30 Bulgarian forces began the Second Balkan War by launching a surprise assault on Greek and Serbian positions in Macedonia. As

Second Balkan War

the Bulgarian attack was being repulsed, Romanian troops began an uncontested march toward Sofia from the north, and Turkey reoccupied the fortress of Adrianople. By the Treaty of Bucharest (Aug. 10, 1913), Romania took the rich lands of the southern Dobruja and the city of Silistra, while Serbia and Greece divided the larger part of Macedonia between them. From its gains in the First Balkan War, Bulgaria retained only a small part of eastern Macedonia, the Pirin region, and a portion of eastern Thrace. This was poor compensation for the loss of the southern Dobruja and of the Bulgarian exarchate in Macedonia. Consequently, the desire to win back what had been lost was the main motivating factor in Bulgaria's diplomacy when World War I began.

WORLD WAR I

Bulgaria's alliance with the Central Powers. When World War I began, Bulgaria declared strict neutrality, but the Tsar and a Germanophile government under Vasil Radoslavov encouraged both sides to bid for Bulgarian intervention. In this contest, the Central Powers (Austria-Hungary and the German Empire) could offer far more at the expense of Serbia, Greece, and, later, Romania, than could the Triple Entente (an alliance of Great Britain, France, and Russia), which had to take the interests of its smaller allies into account. During the summer of 1915, when the military balance swung in Germany's favour, Bulgaria committed to the Central Powers, and Bulgaria declared war on Serbia on October 14. Some of the neutralist and pro-Entente political figures objected, but none went as far as the Agrarian leader Stamboliyski, who threatened the Tsar and issued a call for the troops to resist mobilization. For these acts he was arrested and condemned to life imprisonment.

Bulgaria's losses. By the autumn of 1918 approximately 900,000 men, nearly 40 percent of the male population, had been conscripted. The army suffered 300,000 casualties, including 100,000 killed, the most severe per capita losses of any country involved in the war. In the interior, bad weather and the absence of adult male labour cut grain production nearly in half, while the town population suffered from shortages of food and fuel and from runaway inflation. "Women's riots" for food began early in 1917 and continued to the end of the war. The revolutions in Russia and the hopes inspired by American intervention in the war and by U.S. president Woodrow Wilson's Fourteen Points peace plan further contributed to the breakdown of civilian order and military discipline. In June 1918 the replacement of the pro-German Radoslavov by Alexander Malinov, a leader of the parliamentary opposition, raised hopes for an end to the war, but instead frustration increased as Malinov yielded to Tsar Ferdinand's determination to fight on.

Domestic disorder. On Sept. 15, 1918, the Allied forces on the Macedonian front broke through the Bulgarian lines at Dobropol'ye. The army dissolved as many of the troops deserted to return home while others began a march on Sofia to punish the Tsar and party leaders responsible for the war. Ferdinand turned to Stamboliyski, releasing the Agrarian leader from prison in return for his promise to use his influence to restore order among the troops. Stamboliyski, however, joined the uprising and, at the village of Radomir, where rebel troops were encamped, proclaimed Bulgaria a republic. The Radomir Rebellion was short-lived as the Agrarian-led assault on Sofia was repulsed by German and Macedonian forces that remained loyal to the Tsar. But this provided only a temporary respite. The Bulgarian government asked the Allies for an armistice, which was signed on September 29. Four days later Tsar Ferdinand abdicated in favour of his son Boris III and left the country.

Consequences of the war. Bulgaria was punished for its part in World War I by the Treaty of Neuilly, which assigned the southern portion of the Dobruja region to Romania, a strip of western territory including Tsaribrod and Strumitsa to the Kingdom of Serbs, Croats, and Slovenes (modern Yugoslavia), and the Aegean territories gained in the Balkan Wars to the Allies, who turned them over to Greece at the Conference of San Remo in 1920. Bulgaria was also disarmed and subjected to a heavy burden of reparations.

Postwar politics and government. Defeat and the hardships of war broke the hold of Bulgaria's traditional parties on the government. In the first two postwar elections the Agrarians, Communists, and Socialists together polled first 59 percent and then 65 percent of the ballots. These parties were not united, however, and a Communist-led general strike in the winter of 1919–20 was ruthlessly put down by Stamboliyski, who became prime minister first in coalition with smaller conservative parties and then as head of an all-Agrarian cabinet.

Programs of the Agrarian Union. The years from 1920 to 1923 represented a remarkable period in which the Agrarian Union sought to translate into reality the beliefs and ideas developed in its years in opposition. The Agrarian government introduced a progressive income tax and a land reform directed against the country's few large estates and against absentee ownership, sponsored the spread of cooperative organizations in agriculture and other branches of the economy, and undertook a massive expansion of the school system, providing, among other things, for free, compulsory secondary education. The Agrarians also introduced the practice of compulsory labour service, by which all young men were required to contribute a year's labour on state projects in lieu of military conscription.

Stamboliyski's foreign policy. Stamboliyski abandoned the traditional Bulgarian goal of territorial expansion, which had required huge military budgets, a standing army and professional officer corps, and the patronage of the great powers. His policy aimed, above all, to cultivate good relations with the Kingdom of Serbs, Croats, and Slovenes by accepting the status quo in Macedonia.

Stamboliyski's policies alienated the old political leaders, the Military League (composed of active and reserve officers), and Tsar Boris' court. Influenced by the successes of Benito Mussolini's Fascist organization (known as the Blackshirts) in Italy, the rightist parties united in the National Alliance (later called Democratic Alliance) and planned to march on Sofia to wrest control of the country. On the left, the Communists viewed the Agrarian government as their principal opponent. But the most dangerous enemies were the Military League and the Internal Macedonian Revolutionary Organization (IMRO). IMRO established effective control over the Pirin region and launched terrorist attacks across the border into Yugoslav and Greek Macedonia. It also assassinated several Agrarian leaders. Unable to rely on the Bulgarian military against the Macedonian terrorists, Stamboliyski turned to Yugoslavia; by signing the Treaty of Niš he permitted Yugoslav forces to pursue the Macedonian bands into Bulgarian territory. This treaty, combined with an overwhelming Agrarian election victory in early 1923, led Stamboliyski's opponents to plan a coup d'état. The coup was organized by the Military League, IMRO, and the old parties, and it probably had the support of Tsar Boris III. When it was launched on the night of June 8–9, 1923, it took the Agrarian government by surprise. Stamboliyski was captured a few days later and brutally murdered, and a right-wing government under Aleksandŭr Tsankov took his place.

Stamboliyski's overthrow

The Communist Uprising. The Bulgarian Communists, who had declared their neutrality when the coup occurred, were chastised by Moscow and directed to prepare an armed revolt against the Tsankov regime. The Communists' September Uprising was put down with great brutality and provided Tsankov with a pretext for outlawing the party. The Communists struck back in 1925 with an attempt to assassinate the Tsar and leaders of the government by blowing up Sofia's Sveta Nedelya Cathedral during services. Although 123 people were killed and hundreds more wounded, the main targets escaped, and the government exacted brutal reprisals. As a result of the defeats suffered in 1923 and 1925, many of Bulgaria's Communist leaders escaped abroad, finding positions in the Soviet Union or the Comintern (Communist International), and one, Georgi Mikhailovich Dimitrov, achieved international fame as the chief defendant in the Reichstag

Fire trial of 1933. So great was his standing that, following his acquittal, Stalin had him appointed secretary general of the Comintern, a position he held until that body was dissolved in 1943.

Attempts to stabilize government. After 1925 Bulgarian political life began a slow recovery. In January 1926 Tsankov yielded the premiership to the more moderate Andrei Liapchev. A gradual and qualified return to a free press and parliamentary politics marked his five-year tenure, although assassinations and other terrorist acts by IMRO continued and soured Bulgaria's relations with Yugoslavia and Greece. In 1931 a reconstituted opposition called the Popular Bloc, a coalition that included the Agrarian Union, defeated the Democratic Alliance and pledged to restore full democracy.

The Popular Bloc government came to power during a period of economic crisis brought on by the world depression. Its inability to alleviate the economic situation led to further political instability and a rising tide of labour unrest. On the night of May 18–19, 1934, the Military League carried out a peaceful coup d'état that installed Kimon Georgiev as prime minister. The ideology of the new regime was supplied by an elitist group called Zveno ("a link in a chain"), which drew its membership from intellectual, commercial, and military circles. Inspired by the former Socialist Dimo Kazasov, Zveno advocated "national restoration" through an authoritarian, technocratic regime. The "divisive forces" associated with parliamentary politics were eliminated by the suspension of the constitution and the suppression of all political parties. A new assembly was created, composed of individuals without party affiliation and elected from approved government lists.

The new regime was able to suppress IMRO gangsterism and restore the government's authority over Pirin Macedonia, but its political base was too narrow to allow it to consolidate power firmly. The real beneficiary of the 1934 coup was Tsar Boris III. By the end of 1935 he had relied on his own clique in the army to unseat and jail Georgiev, purge the Military League, and install a subservient government under Georgi K'oseivanov. Bulgaria thus ended the 1930s as a royal-military dictatorship, the form of government that had become nearly universal in eastern Europe.

WORLD WAR II

Boris' alliance with Germany. When World War II began in Europe, Bulgaria proclaimed neutrality. Tsar Boris, however, appointed a new government under a notorious Germanophile, Bogdan Filov, and moved steadily closer to the German orbit, particularly after Germany forced Romania to restore the southern Dobruja to Bulgaria in August 1940. The desire for territorial expansion at the expense of Yugoslavia and Greece and the expectation of a German victory led Boris to join the Axis on March 1, 1941. German troops used Bulgaria as a base from which to attack Yugoslavia and Greece, and in return Bulgarian forces were permitted to occupy Greek Thrace, Yugoslav Macedonia, and part of Serbia. After the German invasion of the Soviet Union and the Japanese attack on the U.S. naval installation at Pearl Harbor, Bulgaria yielded to German pressure to declare war on Great Britain and the United States, a move that was regarded as of only symbolic importance, but Tsar Boris avoided joining the war against the Soviet Union, fearing that this would lead to popular unrest.

Bulgaria was relatively untouched by military operations until the summer of 1943, when Allied bombers began to attack rail and industrial centres. The government complied with German requests to deport the Jews of the occupied territories to the concentration camp at Treblinka, but Tsar Boris cancelled orders for the deportation of the rest of Bulgaria's Jews in the face of protests led by prominent political figures and Metropolitan Stefan, head of the Bulgarian Orthodox Church.

Bulgarian resistance to the Axis alliance. Some attempts at forming a resistance were made by Agrarian leaders when Bulgaria joined the Axis. After the German attack on the Soviet Union, however, it was the Bulgarian Communist Party that assumed the initiative inside the country. Until the final stage of the war, the tactics of the resistance emphasized sabotage and small-group operations. About 10,000 persons are estimated to have participated, making it the largest resistance movement among any of Germany's allies. Politically, the Communists sought the cooperation of other opposition groups, and in August 1943 the Fatherland Front was formed, composed of Communists, left-Agrarians, Zveno, the Socialists, and some independent political figures. The Front's influence grew as the military situation of Germany deteriorated.

Many Bulgarians expected Tsar Boris to break with the German alliance when circumstances permitted. On Aug. 28, 1943, however, the Tsar suffered a fatal heart attack. Because his son and heir, Simeon II, was only six years old, Filov established a Regency Council headed by himself and appointed a new government under Dobri Bozhilov that remained loyal to the German alliance. In May 1944, faced with the continuing German collapse and stern Allied threats that Germany's allies would be severely punished, Bozhilov resigned. He was replaced with the right-Agrarian Ivan Bagrianov, who began secret negotiations for surrender with the Allies but at a snail's pace. At the end of August, the sudden surrender of Romania, which brought Soviet troops to the Danube months before they had been expected, created panic in Sofia. When Bagrianov's attempt to proclaim Bulgarian neutrality was rejected as insufficient by both Britain and the Soviet Union, the prime minister resigned and was replaced by Kosta Muraviev of the Agrarian Union, who reiterated a proclamation of neutrality. Four days later, on Sept. 5, 1944, the Soviet Union declared war on Bulgaria, and Soviet forces began to enter the country unopposed. Simultaneously, the Fatherland Front began preparations for a coup d'état. Despite Muraviev's declaration of war against Germany on September 8, during that night military forces organized by Zveno and partisan detachments occupied key points in Sofia and toppled Muraviev's government in the name of the Fatherland Front. Kimon Georgiev of Zveno became the new prime minister and sought an immediate armistice with the Soviet command.

THE POSTWAR ERA

Consolidation of Communist power. The consolidation of Communist power in Bulgaria was carried out by 1948, coinciding with the completion of the peace treaty with the Allies and the presence of Soviet occupation forces. In the coalition Fatherland Front government, the Communists enjoyed control of the interior and judicial ministries, which were crucial in setting up the new state. Taking advantage of the popular feeling that those who were responsible for Bulgaria's involvement in the war should be punished, the regime established "people's courts" to prosecute the political leaders of the wartime period. The first mass trial, lasting from Dec. 20, 1944, to Feb. 1, 1945, resulted in death sentences for three of the Tsar's former regents, 28 former ministers, 68 members of parliament, and an adviser to the Tsar. Other defendants received long prison sentences. According to official statistics, by the time the people's courts completed their work in April 1945, they had tried 11,122 people, of which 2,730 were condemned to death, 1,305 to life imprisonment, and 5,119 to terms of up to 20 years. Unofficial estimates suggested that as many as 30,000 political opponents were killed.

The army, which was under overall Soviet command but was not yet under direct Communist control, was engaged in the final stages of the campaign against Germany. The Communists insisted on the immediate appointment of officers drawn from the partisan movement, and by the time the regular army returned, the party was in a position to carry through a purge of the officer corps.

Return of Dimitrov from exile

On Nov. 4, 1945, Georgi Dimitrov returned to Bulgaria after 22 years of exile. The belligerent tone with which he referred to the opposition in his first public address was a clear indication of his party's intentions.

Given the Communist Party's control of the instruments of power, the hopes of the non-Communist opposition rested on the Western democracies. Indeed, during the summer of 1945 the regime postponed parliamentary elections after Great Britain and the United States protested the undemocratic character of the proposed electoral laws. Bulgaria, however, did not have a high priority on the diplomatic agenda of the West. As early as October 1944 British prime minister Winston Churchill had shown his willingness to consign the country to Soviet control during his "percentages discussion" with Soviet premier Joseph Stalin. As the Bulgarian Communists and their Soviet sponsors became aware that the Western states were not concerned with events in Bulgaria, they moved more forcefully to eliminate the internal opposition. Elections held in November 1945 returned a substantial majority in favour of the Communists and their allies. In September 1946 a referendum decided by a 93-percent majority to proclaim Bulgaria a republic, and Tsar Simeon II and the Queen Mother were required to leave the country.

Elections of 1946

Elections for a Grand National Assembly to prepare a new constitution were held on Oct. 27, 1946. Although they produced a large Communist majority, the opposition polled more than one million votes, which amounted to 28 percent of the total. When the Assembly opened in November, the Agrarian leader Nikola Petkov emerged as the opposition's principal spokesman. He was charged by the police with plotting to overthrow the government and was expelled from parliament along with most of his associates. In June 1947 Petkov was arrested, and on September 23 he was executed. One week later the United States extended diplomatic recognition to the new regime; Great Britain had already done so in February.

The defeat of the political opposition coincided with or was followed by the elimination of nearly all elements of pluralism in Bulgarian society. This seems to have been accelerated after the founding congress of the Communist Information Bureau (Cominform) on Sept. 22–27, 1947, at Szklarska Poredba in Poland, where Andrey Aleksandrovich Zhdanov and Georgy Maksimilianovich Malenkov delivered the message that Stalin desired a more rapid transformation of the Socialist camp along Soviet lines. In Bulgaria the "intensification of the revolutionary process" saw increased pressure on the remaining non-Communist parties. Some Socialists applied to join the Bulgarian Communist Party, and those who remained in opposition were crushed by police repression. The Socialist Party was formally absorbed by the Communists in August 1948. The Agrarian leader, Georgi Traikov, repudiated his party's traditional ideology and defined a new role for it as the helpmeet or "little brother" of the Bulgarian Communist Party in the countryside. Zveno and the remaining smaller parties announced their "self-liquidation" by the beginning of 1949. They were dissolved into the Fatherland Front, which in turn was converted into a broad patriotic organization under Communist control.

The "Dimitrov Constitution"

In the Grand National Assembly a team of Soviet jurists assisted in the preparation of the "Dimitrov Constitution," which went into effect on Dec. 4, 1947. Modeled closely on the Soviet constitution of 1936, it provided a legal foundation for the reconstruction of the state on Communist principles.

Reforms under the Fatherland Front. The Fatherland Front regime had launched an assault on private property almost immediately after the coup of Sept. 9, 1944, employing a variety of legislative measures aimed at confiscating the wealth of "fascists" or "speculators." The Dimitrov Constitution provided for even larger measures of nationalization. All large-scale industries, banks, and insurance companies were nationalized, and government monopolies were established over the major items of retail trade. By the end of 1948 approximately 85 percent of industrial production was in the hands of the state, with another 7 percent carried on by cooperative organizations. The party also sponsored a reorganization of the country's trade unions, creating the General Workers' Trade Union and gradually forcing all workers' organizations into it. Similarly, the youth organizations of the various parties were incorporated into the Dimitrov Communist Youth League.

Exarch Stefan, head of the Bulgarian Orthodox Church, sought to adapt to the new political regime, but he resisted the efforts of the Bulgarian Communist Party to control church affairs directly. In September 1948 Stefan resigned his office under mysterious circumstances and retired to a monastery. His successor offered no resistance to legislation adopted in March 1949 that subjected all religious orders to state supervision. At the time this law was adopted, 15 pastors from evangelical Protestant churches were arrested, tried, and executed for espionage on behalf of the United States and Britain and for other alleged crimes. Soon afterward a number of Bulgarian Catholic clergy were tried for spying for the Vatican and for disseminating anti-Communist propaganda. The nearly 50,000 Bulgarian Jews who survived the war were encouraged to emigrate to Israel. The regime also attempted to deport ethnic Turks and Gypsies, causing the Turkish government to seal the border.

Stalinism and de-Stalinization. Traicho Kostov, who had been instrumental in supervising the defeat of the opposition, was accused of treason and of collaborating with Yugoslavia's Communist leader Josip Broz Tito against Stalinism. Kostov's execution in December 1949 was followed by the purge of thousands of "Kostovites" and others alleged to be criminals and spies.

Economic developments

Georgi Dimitrov died in office in July 1949. His immediate successor, Vasil Kolarov, died early in 1950 and was followed by Vulko Chervenkov. Known as Bulgaria's "Little Stalin," Chervenkov followed policies aimed at developing Bulgaria according to the Soviet model. These included rapid industrialization, with an emphasis on heavy industry, the forced collectivization of agriculture, heavy reliance on the police and security apparatus to control the country, and isolation from nations outside the Soviet bloc.

Stalin's death in 1953 and the inauguration of the "New Course" in the Soviet Union had repercussions in Bulgaria. In 1954 Chervenkov accepted the Soviet model of collective leadership, remaining prime minister but yielding his post as party leader to Todor Zhivkov. The government also released several thousand political prisoners and moderated its economic policies in favour of raising living standards. The beginning of open de-Stalinization at the Soviet Union's 20th Communist Party Congress in February 1956 was followed in Bulgaria by the April Plenum of the Bulgarian Communist Party, at which Chervenkov was accused of having created a personality cult and of having violated the norms of legality. Chervenkov was removed from the premiership, there was some relaxation of censorship, and the victims of the Kostovite trials, including Kostov himself (posthumously), began to be rehabilitated.

Rise of Todor Zhivkov. In the years after the April Plenum, party leader Zhivkov became the dominant figure in political life. In 1962 he became prime minister, and he continued to hold the positions of head of state and head of party until 1989. An attempted putsch led by General Ivan Todorov-Gorunya in 1965 was easily put down, and Zhivkov consistently managed to purge or undercut party leaders regarded as potential rivals.

During the era of Zhivkov's ascendancy, Bulgaria modeled its domestic policies on those of the Soviet Union. Treaties linked Bulgaria's economic development with that of the Soviet Union through the end of the 20th century. Bulgaria gave the highest priority to participation in the modern scientific-technological revolution and pursued policies aimed at industrialization and the development of a population with the education and skills appropriate to an industrial state. In 1948 approximately 80 percent of the population drew their living from the soil. In 1988 the government reported that 19 percent of the labour force was engaged in agriculture, with the rest concentrated in industry and the service sector.

In its foreign relations after the 1950s, Bulgaria abandoned the isolationism that characterized the Chervenkov period. Although remaining steadfast in its commitments to the Warsaw Pact and the Council for Mutual Economic Assistance, Bulgaria improved relations with its Balkan neighbours, particularly Greece, and expanded its economic and cultural relations with most Western states. Relations with Yugoslavia remained strained, however, over the persistence of the Macedonian issue. In 1979

Bulgaria proposed a treaty with Yugoslavia that would guarantee the inviolability of the borders established after World War II; this proposal was rejected, however, because of Bulgaria's refusal to admit the existence of a distinct Macedonian nationality. From the Bulgarian point of view, such an admission would both fly in the face of historical reality and legitimize Yugoslav claims on the Pirin region.

During the 1970s, in spite of the country's economic growth, a serious concern appeared about the low birth rate of the ethnic Bulgarian population. Numerous measures were adopted to encourage larger families, but without apparent effect. In late 1984 the government began a major effort to convert the country's ethnic Turks into Bulgarians. Measures aimed at the Turkish population, estimated to number approximately 800,000, included the abandonment of Turkish-language publications and radio broadcasts and the requirement that Turks adopt Bulgarian names. Despite these policies, the ethnic Turkish population was not assimilated, and clashes with the authorities continued. During the spring and summer of 1989, when the government of Turkey offered to accept refugees from Bulgaria, more than 300,000 ethnic Turks fled or were driven from the country in a wave of violence that focused worldwide attention on Bulgaria's violations of human rights.

The era of reforms launched by Mikhail Gorbachev in the Soviet Union had a major impact on Bulgaria, inspiring greater demands for openness and democratization. The growth and greater aggressiveness of Bulgarian dissidents, a declining economic situation, and internal party rivalries led Zhivkov's colleagues to force his resignation on Nov. 10, 1989. He was soon subjected to intense criticism for corruption and abuse of power, placed under detention, and held while a state commission prepared formal charges against him.

Post-Zhivkov reforms. Zhivkov's successors endorsed a policy of openness, pluralism, and respect for law, halted repression of the ethnic Turks, and took the first steps toward separating the Communist Party from the state. Article One of the constitution, guaranteeing the party a monopoly of power, was repealed. After some shuffling of positions, Petar Mladenov was named head of state, Andrey Lukanov prime minister, and Alexander Lilov head of the Communist Party. In early 1990 the party held an extraordinary congress that enacted significant changes in party structure and purged many of the old guard. To symbolize its break with the repressive policies of the past, it abandoned the name Communist and renamed itself the Bulgarian Socialist Party.

Bulgarian Socialist Party

In the meantime, dissident groups took advantage of the country's new freedoms to organize opposition political parties. Many of these joined the Union of Democratic Forces (UDF), a coalition led by the sociologist Zheliu Zhelev. By the spring of 1990 the UDF and the Socialists came to an agreement on the holding of free elections for a Grand National Assembly that would prepare a new constitution for the country. In these elections and subsequent runoffs, held June 10 and June 17, the Socialists won 47 percent of the vote and a narrow majority (211 of 400 seats) in the Assembly. In July 1990 Mladenov resigned after unsuccessfully attempting to conceal the fact that he had recommended a military crackdown on protesters in late 1989. Because their majority was too small to allow them to govern alone, the Socialists supported the election of opposition leader Zhelev in August. (J.D.B.)

For later developments in the history of Bulgaria, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 921, 924, 962, 963, and 972, and the *Index.*

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Business Law

Business law—also called commercial law, or mercantile law—is not recognized specifically as a branch of common law but, in popular usage, is frequently an omnibus term given to titles of textbooks and academic courses on such commercial subjects as contract, agency, commercial transactions, bankruptcy, and labour. In most continental European legal systems based on Roman law, however, "commercial law," or "mercantile law," designates a branch of law distinct from "civil law"—the commercial law governing merchants or businesses engaged in acts of commerce and the civil law governing ordinary consumer transactions or other acts of persons not engaged in trade. In continental European law, the rules for the two branches differ.

The sections below deal with major branches of law relevant to business and commercial practices. The law of business partnerships and limited-liability companies, or corporations, is covered in BUSINESS ORGANIZATION. Allied branches of law may be found under PROPERTY LAW; additional aspects of labour law are treated in the article WORK AND EMPLOYMENT.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 553.

This article is divided into the following sections:

Contracts

The nature of contract law

A contract, in the simplest definition, is a promise enforceable by law. The promise may be to do something or to refrain from doing something. The making of a contract requires the mutual assent of two or more persons, one of them ordinarily making an offer and another accepting. If one of the parties fails to keep the promise, the other is entitled to legal recourse against him. The law of contracts considers such questions as whether a contract exists, what the meaning of it is, whether a contract has been broken, and what compensation is due the injured party.

HISTORICAL DEVELOPMENT

Contract law is the product of a business civilization. It will not be found, in any significant degree, in precommercial societies. Most primitive societies have other ways of enforcing the commitments of individuals; for example, through ties of kinship or by the authority of religion. In an economy based on barter, most transactions are self-enforcing because the transaction is complete on both sides at the same moment. Problems may arise if the goods exchanged are later found to be defective, but these problems will be handled through property law—with its penalties for taking or spoiling the property of another—rather than through contract law.

Even when transactions do not take the form of barter, primitive societies continue to work with notions of property rather than of promise. In early forms of credit transactions, kinship ties secured the debt, as when a tribe or a community gave hostages until the debt was paid. Other forms of security took the form of pledging land or pawning an individual into "debt slavery." Some credit arrangements were essentially self-enforcing: livestock, for example, might be entrusted to a caretaker who received for his services a fixed percentage of the offspring. In other cases—constructing a hut, clearing a field, or building a boat—enforcement of the promise to pay was more difficult but still was based on concepts of property. In other words, the claim for payment was based not on the existence of a bargain or promise but on the unjust detention of another's money or goods. When a worker sought to obtain his wages, the tendency was to argue in terms of his right to the product of his labour.

A true law of contracts—that is, of enforceable promises—implies the development of a market economy. Where a commitment's value is not seen to vary with time, ideas of property and injury are adequate and there will be no enforcement of an agreement if neither party has performed, since in property terms no wrong has been done. In a market economy, on the other hand, a person may seek a commitment today to guard against a change in value tomorrow; the person obtaining such a commitment feels harmed by a failure to honour it to the extent that the market value differs from the agreed price.

Roman law. The Roman law of contracts, as found in Justinian's law books of the 6th century AD, reflected a long economic, social, and legal evolution. It recognized various types of contracts and agreements, some of them enforceable, others not. A good deal of legal history turns upon the classifications and distinctions of the Roman law. Only at its final stage of development did Roman law enforce, in general terms, informal executory contracts—that is, agreements to be carried out after they were made. This stage of development was lost with the breakup of the empire. As western Europe declined from an urbanized, commercial society into a localized, agrarian society, the Roman courts and administrators were replaced by relatively weak and imperfect institutions.

The rebirth and development of contract law was a part of the economic, political, and intellectual renaissance of western Europe. It was everywhere accompanied by a commercial revival and the rise of national authority. Both in England and on the Continent, the customary arrangements were found to be unsuited to the commercial and industrial societies that were emerging. The informal agreement, so necessary for trade and commerce in market economies, was not enforceable at law. The economic life of England and the Continent flowed, even after a trading economy began to develop, within the legal framework of the formal contract and of the half-executed transaction (that is, a transaction already fully performed on one side). Neither in continental Europe nor in England was the task of developing a law of contracts an easy one. Ultimately, both legal systems succeeded in producing what was needed: a body of contract doctrine by which ordinary business agreements, involving a future exchange of values, could be made enforceable.

The law merchant

The new contract law began to grow up throughout Europe through the practices of merchants; these were at first outside the legal order and could not be upheld in courts of law. Merchants developed informal and flexible practices appropriate for active commercial life. By the 13th century, merchants' courts had been established at the international trade fairs. The merchant courts provided expeditious procedures and prompt justice and were administered by men who were themselves merchants and thus fully aware of mercantile problems and customs.

In the 12th and 13th centuries the development of the law of contracts on the Continent and in England began to diverge. In England the common law of contracts developed pragmatically through the courts. On the Continent the process was very different, with speculative and systematic thinkers playing a much larger role.

Common law. From perhaps the 13th century on, English common law dealt with contractual problems primarily through two actions: debt and covenant. When a fixed sum of money was owed, under an express or implied agreement, for a thing or a benefit given, the money was recoverable through a simple action at debt. Other debt action was available for breach of a promise, made in an instrument with a seal, to pay a fixed sum of money. A so-called action at covenant could also be brought, but only for breach of a promise under seal. These actions did not, however, provide a remedy for the breach of an informal agreement to do something. In the 15th century the common-law courts started to develop a form of action that would render such agreements enforceable, and by the middle of the 16th century they had done so through the form of action known as assumpsit ("he has undertaken"). Originating as a form of recovery for the negligent performance of an undertaking, it came step by step to cover the many kinds of agreement called for by expanding commerce and technology. Having established in principle a comprehensive remedy, it was necessary for the courts to limit its scope. The courts found the limiting principle in the doctrine of "consideration," according to which a promise as a general rule is not binding unless something is given or promised in exchange. This consideration need not be of commensurate value, but it must be of some value, must be bargained for, and cannot be simply a formality.

Civil law. On the Continent, the revived study of classical Roman law had an immense influence upon the developing law of contract. It stimulated the rediscovery or construction of a general law concerning the validity of agreements. The Roman law, however, as crystallized in Justinian's law books, tended to confirm the notion that something more than an informal expression of agreement was required if a contract was to be upheld by a court. Another significant influence in the development of contract law on the Continent was the Roman Catholic Church. The church in its own law (canon law) strongly supported the proposition that a simple, informal promise should be binding (*pacta sunt servanda*). This attitude was to encourage the development of informal contracts. The natural-law philosophers took up such ideas as *pacta sunt servanda,* although they were slow to abandon the view that some contracts, especially contracts of exchange, should require part performance if they were to be held enforceable. By the 18th century the speculative and systematic thought of jurists and philosophers had finally and fully carried the day. The legal writers and legislators of the period generally considered informal contracts as enforceable in the courts. Thus in the French Civil Code of 1804, contract was approached essentially in terms of agreement; obligations freely assumed were enforceable except when the welfare of society or the need to protect certain categories of persons, such as minors, dictated otherwise. With the generalization that contract rests ultimately on agreement, the civil-law systems achieved a foundation quite different from the common law's view that contract is basically a promise supported by a consideration.

Flexibility of modern contracts

All the Western systems of modern contract law provide mechanisms through which individuals can voluntarily assume, vis-à-vis others, legally binding obligations enforceable by the other person. Contract law strives to give legal expression to the endlessly varying desires and purposes that human beings seek to express and forward by assuming legal obligations. The resulting system is open-ended; in principle, no limits are set in modern contract law to the number of possible variations of contracts.

THE SETTING OF STANDARDS

In theory, contractual obligations should be concluded between parties of substantially equal awareness and bargaining power and for purposes fully approved by society. The law reflects this utopian idea in the sense that it tends to conceive of contract as an arrangement freely negotiated between two or more parties of relatively equal bargaining power. The manifestations of intention required to form a contract are accordingly thought of as indicating real willingness, although in fact they may simply represent acquiescence.

Fairness and social utility. Much of the law of contract is concerned with ensuring that agreements are arrived at in a way that meets at least minimum standards respecting both parties' understanding of, and freedom to decide whether to enter into, the transactions. Such provisions include rules that void contracts made under duress or that are unconscionable bargains; protection for minors and incompetents; and formal requirements protecting against the ill-considered assumption of obligation. Thus, section 138 of the German Civil Code renders void any contract "whereby a person profiting from the distress, irresponsibility, or inexperience of another" obtains a disproportionately advantageous bargain. In addition, more general social requirements and views impinge upon contracts in a number of ways. Certain agreements are illegal, such as—in the United States—agreements in restraint of trade. Others, such as an agreement to commit a civil wrong, are held by the courts to be contrary to the public interest. Certain systems discourage some purposes, such as the assumption of a legally binding obligation to confer a gift of money or other gratuitous benefit upon another, by various special requirements.

Legal systems often have recourse to interpretation in the interest of fairness and social utility. Many litigated cases in which a remedy is sought for breach of contract are concerned with the meaning to be attached to the verbal expressions and acts of the parties in their dealing with each other. Ambiguities, for example, may be resolved against the party thought to have the superior bargaining position. This decision is common in cases in which one party is able to set the terms of a contract without bargaining. Again, a written agreement may be interpreted against the party who drafts or chooses the language. Or the court may prefer an interpretation it finds to be in accord with the public interest.

Although all legal systems try to achieve a reasonable approach to freedom of contract, there are bound to be contractual obligations that depart in some degree from the ideal. No one seeking to enforce a contract is required to show affirmatively that it advances specific ends desired by society or that the contracting process is without blemish. Such a requirement would be administratively cumbersome and expensive. In addition, it would reduce the general usefulness of the contract as an economic and social instrument. Differences in the economic resources available to individuals are found in most societies; to the extent that these differences flow from general conditions and are reflected in, rather than produced by, individual contracts, it is usually not feasible to take remedial action through the law of contracts. A single contract, moreover, is often only one element in a complex of economic and legal relations. Thus, in times of severe inflation or deflation, it may simply not be feasible to seek to deal with the resulting inequities in terms of redoing individual contracts.

Big industry and the public interest

Contracts of adhesion. There are large areas of economic life in which the parties to contracts have such unequal bargaining positions that little real negotiation takes place. These contracts are often known as contracts of adhesion. Familiar examples of adhesion contracts are contracts for transportation or service concluded with public carriers and utilities and contracts of large corporations with their suppliers, dealers, and customers. In such circumstances a contract becomes a kind of private legislation, in the sense that the stronger party to a large extent assigns risks and allocates resources by its fiat rather than through a reciprocal process of bargaining. Enforcement of such standard contracts can be justified on the ground that they are economically necessary. The question then becomes whether these decisions are to be made by private enterprise or by other agencies of society—in particular, government—and to what extent the interest of those who deal with such economic enterprises can be represented and protected in the decision-making process.

Contract law in such cases provides only what can be called the legal relationship. The content of the relationship derives not from bargaining between the parties but from the fiat of the large enterprise often offset by the fiat of some government agency. In a sense, the socially regulated contract of adhesion seeks to eat the cake of bureaucratic rationality while having, as well, the cake of individual choice and decision. Doubtless both cakes are diminished in the process, but the result may well be more satisfying than if only one had to be chosen. At all events, the resulting legal-economic phenomenon is radically different from that envisaged by traditional contract law. In the late 20th century legislative attempts have been made in a number of countries, such as West Germany (1976), the United Kingdom (1977), and France (1978), to strike a balance between the general freedom to contract and the protection of the weaker party.

THE RULES OF DIFFERENT LEGAL SYSTEMS

Traditional contract law developed rules and principles controlling the voluntary assumption of obligations, regulating the performance of obligations so assumed, and providing sanctions for failure to perform.

The question of whether a contract exists

Offer and acceptance. Some of the rules respecting offer and acceptance are designed to operate only when a contrary intention has not been indicated. Thus, in German law an offeror cannot withdraw his offer until the time stipulated in the offer or, if no time is stipulated, until a reasonable time has passed; but this rule yields to a statement in the offer to the effect that it shall be revocable. In Anglo-American common law, when parties contract by correspondence the acceptance takes place on dispatch of the letter, but the offeror can stipulate that no contract will be formed until the acceptance has reached him. These rules serve to fill in points on which the parties in their negotiations have not, for one reason or another, been specific.

Another function of rules relating to offer and acceptance is to enable the parties to understand and to mark when their discussions pass from an exploratory stage to the stage of commitment. The concepts of offer and acceptance are somewhat formal; they assume that the negotiations pass through clearly distinguishable phases, which is often not the case. But they help the parties to distinguish negotiation from commitment. The two words offer and acceptance become firmly associated with the assumption of obligations.

Different legal systems frequently advance comparable policies in quite different ways. Several distinctly different patterns are found in the approach of modern legal systems to the problems of whether an offeror is free to revoke his offer before acceptance, and of when an acceptance is effective to form a contract. Perhaps the polar extremes are represented by the West German civil law, on one hand, and the Anglo-American common law on the other. In the German view, an offer binds the offeror for any stipulated period or, when the offer is silent as to time, for a reasonable period unless the offeror has expressly made the offer revocable. The common-law rule is the opposite: an offer is revocable until it has been accepted. The two systems also have sharply divergent rules with respect to the point at which, when the parties are contracting by correspondence, the acceptance takes effect to conclude the contract. In German law, the acceptance takes effect when it reaches the offeror, in the sense that he either knows or can learn of it. In the common law, on the other hand, if the offeree uses an appropriate means of communication, the acceptance is effective on dispatch unless the offeror stipulated the contrary in his offer. (A revocation by the offeror, however, does not take effect until received by the offeree.)

How are these divergencies in the rules respecting offer and acceptance to be explained? In particular, do they reflect fundamental policy differences or simply different techniques designed to forward quite similar purposes? An examination of a typical problem posed when parties contract by correspondence suggests the latter explanation. Upon receipt of an offer, the offeree frequently changes his position by, for example, refusing or ignoring other offers, neglecting to seek additional offers, or himself making propositions based on the offer made to him. For this reason the legal system sees a need to provide the offeree with a secure point of departure for his decision, in order both to protect him and to facilitate commerce generally. The German system provides this protection by making the offer in principle irrevocable. The common law, on the other hand, found this solution excluded by its doctrine of consideration; as the offeree does not give anything in exchange for the offer's irrevocability, consideration is lacking to support an obligation not to revoke. (On the other hand, the Uniform Commercial Code, which has been adopted everywhere in the United States except in parts of Louisiana, provides that a firm offer made by a merchant is irrevocable even though the other party has given no consideration.) The common law is not entirely insensitive to the offeree's predicament. The rule that the acceptance is effective upon dispatch creates a situation in which the offeror who wishes to revoke his offer is uncertain whether or not he can still do so, since his revocation is not effective until receipt, whereas the offeree's acceptance, if one is made, takes effect on dispatch. This uncertainty makes the consequences of an attempted revocation unpredictable and thereby inhibits an offeror who might otherwise seek to revoke. In sum, the German and Anglo-American systems both try to achieve, and in a measure succeed in achieving, a fair balance between the offeror and the offeree.

The criteria of a legal contract

Unenforceable transactions. In all systems of contract law, certain classes of transactions are treated as unenforceable by the judicial process because they are thought to involve unusual hazards for a contracting party or to be of marginal social utility. There are, in both civil-law and common-law systems, four kinds of concern that lead the systems to treat certain types of transaction as unenforceable. These four kinds of concern may be called evidentiary, cautionary, channeling, and deterrent. The evidentiary concern springs from the desire to protect both the individual citizen and the courts against manufactured evidence and insufficient proof. The cautionary concern seeks to safeguard the individual against both his own rashness and the importuning of others. The channeling concern seeks to mark off or label obligations that may be enforceable and to direct attention to the problem of the extent and kind of the legal obligation, so that the individual will know the legal significance that his action may have. Finally, the deterrent concern refers to those types of transaction that are discouraged because they are felt to be of doubtful value to society.

Two quite different techniques are used to delineate types of transaction that are unenforceable in their natural, or normal, state. The first proceeds by describing the type in functional or economic terms. The common-law Statute of Frauds enacted by the English Parliament in 1677 provided that the following six kinds of contracts should be unenforceable unless expressed in writing: contracts to sell goods exceeding a certain value; contracts to sell any interest in land; agreements that are not to be performed within a year of their making; agreements upon consideration of marriage; suretyship agreements; and undertakings by an executor or administrator to be surety on a debt of the deceased for which the estate is liable. Civil-law systems typically describe as unenforceable in the absence of an appropriate formality noncommercial contractual obligations exceeding a certain value; mortgages created by contract; noncommercial compromise agreements; marriage contracts; agreements binding a party to transfer all, or a fractional part of, his property; leases to run for more than a year; assumptions of the obligation to stand as surety, at least when the operation is not a commercial one on the surety's part; promise of an annuity; and promises to make gifts.

Another less direct technique for delineating unenforceable types of transaction derives from the common law's doctrine of consideration. It holds transactions unenforceable in the absence of a bargained-for exchange. This class would include, for example, promises to make gifts. The approach tends to be too all-embracing, treating certain types of transaction as suspect when there is little or no practical justification for doing so. It is not clearly demonstrated, for example, that an option agreement made by two businessmen should be handled differently from many other kinds of commercial dealings. A strong argument exists that the common law's handling of commercial options, business compromises, and other business transactions lacking an element of exchange is more a logical deduction from the general doctrine of consideration than an expression of justifiable policy concerns.

Except in cases where the ground for unenforceability is radical, when a given transaction type is considered unenforceable the legal system should prescribe an extrinsic element the addition of which will cure the defect—for example, expressing the agreement in writing, performing it in part, or having a document drawn up with the participation of a legally qualified notary or other public official who holds a special appointment from the state and is charged with handling and recording various types of transactions.

Decline of the use of seal

A complex situation has arisen with respect to the two most generally available extrinsic elements, the seal and the payment of a nominal consideration. Various states of the United States no longer consider the seal as an effective extrinsic element. The seal's decline is rooted in its changed significance in the modern, literate, democratic world. The seal was originally an impression, usually in wax, of a device, or design, representing an individual or a family. In modern times, the courts, with legislative assistance in a fair number of the states of the United States, have recognized easy-going substitutes for the wax seal, such as simple writing presumed to have been made for sufficient consideration or, in special circumstances, parol agreement for valid consideration. The effect has been to render the seal progressively less effective, particularly from the cautionary perspective, and many courts now refuse to accept it as a satisfactory formality.

Nominal consideration is a subtle and ingenious formality. Its essence is the introduction of a contrived element of exchange into the transaction. Thus A, desiring to bind himself to give B $10,000, requests B to promise to give (or to give) A a peppercorn in exchange. B's promise (or performance) is an element, extrinsic to a normal gift promise, introduced by the parties in an effort to render the transaction enforceable (since the law does not treat normal gift promises as enforceable). Common-law courts often accept nominal consideration when used in a business context, such as in an option arrangement or a compromise agreement; its effectiveness is understandably more doubtful in the context of a gift promise, since such a transaction involves greater dangers for one party and is socially more marginal.

Civil-law systems have less need than the common law for a formality such as nominal consideration; they prescribe methods directly in their statutes. Interestingly enough, however, in some civil-law systems an analogous, judicially developed formality has emerged—the disguised donation (*donation déguisée*) of French law, in which the parties cast a gift promise in the form of an onerous transaction, such as a sale. It can be argued that both the nominal considerations and the disguised donation serve at least the cautionary and channeling functions of formalities mentioned above.

Reliance

Another kind of extrinsic element recognized by some courts, especially in the common-law countries, is one party's reliance upon the promise of the other. The fact of reliance argues in favour of enforcement because it indicates that an underlying understanding existed between the parties and because the relying party may suffer as a consequence of his change of position. Some courts will enforce initially suspect transactions when several extrinsic elements are present in combination. A common-law court, for example, may enforce a gift promise in which the element of reliance was present in addition to a seal or a nominal consideration. Other extrinsic elements, either alone or in combination with reliance, a seal, or a nominal consideration, may also render a transaction enforceable. Cases, for example, in which the promisor dies without attempting to revoke a gift promise could be enforced, as distinguished from cases in which the promisor seeks to revoke.

Performance. Contract law seeks to protect parties to an agreement not only by requiring formalities but in many other ways as well. Thus rules respecting deceit, fraud, and undue influence are designed to ensure that contractual obligations are assumed freely and without one party misleading the other. Other rules regulate the modification of ongoing contractual relations with a view to preventing a party with considerable bargaining power from unfairly imposing changes in the contract.

The law also allows contractual relations to be adjusted when they have been thrown out of balance by unforeseen circumstances. The task of adjustment is relatively easy in cases in which both parties made a mistake or in which one party laboured under a mistaken assumption that was, or plainly should have been, known to the other. The problem of mistake becomes more intractable when the error is chargeable to only one party. The solutions reached for such situations are complex and defy general statement.

Catastrophic events such as inflation, political upheaval, or natural disasters may upset the economy of a contract. In the case of natural catastrophes, relief is frequently available under theories of force majeure (action by a superior or irresistible force) and "act of God" (act of nature that is unforeseeable and unpreventable by human intervention). When the unsettling circumstances are economic in their nature, as with severe inflation or deflation, a solution is difficult to find. A party who benefits from inflation in one contractual or economic relation may suffer from it in another. A general readjustment in contracts would be enormously complicated and time-consuming and would interject an undesirable element of uncertainty into economic and business activity. Only under exceptional circumstances—and usually in the form of special legislation—are contractual relations adjusted for the effects of severe economic dislocations.

Sanctions for nonperformance

Failure to perform. Another branch of contract law deals with the sanctions that are made available to a contracting party when the other party fails to perform his contractual obligations. When these sanctions take the form of money damages—as they usually do in practice, even though some civil-law systems have a theoretical preference for specific relief—the system must decide whether the plaintiff is to be put in the same position economically that he would have been in had the contract been performed (expectancy damages) or simply reimbursed for the actual losses, if any, flowing from his reliance on the contract (reliance damages). Reliance damages can, of course, be very large. A subcontractor who fails to deliver parts required for the construction of an ocean liner (or delivers faulty parts) may be responsible for heavy reliance damages resulting from delay in the work or actual damage to the vessel. Legal systems utilize various techniques to limit both reliance and expectancy damages when otherwise they would be unreasonably large.

If a person has agreed to buy an article from a merchant, his refusing to take delivery will not ordinarily produce substantial reliance damages. Delivery costs will have been incurred, but the merchant will presumably not have lost sales elsewhere. In such circumstances, the merchant will seek to recover not his delivery costs but his lost profit—his expectancy damages. The law allows relief on the basis that the expectancy created by an enforceable promise has a current economic value, measured by the economic gain that the party would derive if the particular agreement were performed.

In some circumstances, performance is not measureable in terms of market value—as, for example, when one relative has agreed to sell to another a family painting of sentimental value but of little intrinsic worth. Many legal systems in such a case require specific performance (that is, compliance with the precise terms agreed upon in the contract). The availability of specific relief varies among contemporary legal systems, for reasons that seem more historical and doctrinal than practical.

Multiparty contracts

Other problems of contract law. Many contracts involve more than two persons. The law of contracts provides special rules for regulating claims by multiparty plaintiffs or claims against multiparty defendants, or for determining rights among the parties. Multiparty problems arise in other contexts as well. There is the problem of whether the immediate parties to a contract can enter into an agreement that will confer rights upon a person not an original party to the contract. Probably because the dogmatic structure of contract law was largely formed on the model of the simpler two-party situation, and because the contract for the benefit of third parties did not have great practical importance until such relatively modern developments as the emergence of life insurance, many systems of contract law have encountered difficulty in working out the relationship between the third party and the underlying contract. English law took the view that, as a rule, a person cannot acquire a right on a contract to which he is not a party. Some of the problems posed are difficult to resolve: under what circumstances and to what extent should the third party control the underlying contract when, for example, the original parties desire to rescind or modify it?

Another variation of the party problem is presented by efforts to add or substitute parties to a contract. In the absence of an express regulation of the problem in the basic contract, the law works with the notion of the presumed intention of the contracting parties, based on considerations of fairness and practicality. A contracting party cannot, in principle, assign to another his right under a contract if the assignment would result in a significant change in the burden assumed by the other contracting party. A contractual right to receive money or goods is a different matter; it can ordinarily be assigned because the resulting burden on the person under obligation is not great, and because society as a whole benefits from having this flexible economic and legal instrument.

One problem of contract law that has been mentioned above deserves further consideration—the problem of interpretation. Many rules of contract law are simply presumptions, based on experience and tradition, as to what the parties ordinarily intend; if they clearly intend otherwise, the rules are not mandatory. Problems of interpretation frequently arise with respect to the particulars of a given agreement; thus the court seeks to determine what the parties actually had in mind. The effort to ascertain intention may encounter difficulties arising from the law of evidence. Many legal systems limit the use of testimonial evidence to explain the essential elements of a written contract.

CONTEMPORARY TENDENCIES

Arbitration. Modern commercial practice relies to a growing extent on arbitration to handle disputes, especially those that arise in international transactions. There are several reasons for the growing use of arbitration. The procedure is simple, it is more expeditious, and it may be less expensive than traditional litigation. The arbitrators are frequently selected by a trade association or business group for their expert understanding of the issues in the dispute. The proceedings are private, which is advantageous when the case involves trade or business secrets. In many legal systems, the parties can authorize arbitrators to base their decision on equitable considerations that the law excludes. Finally, when the parties are from different countries, an international panel of arbitrators may offer a greater guarantee of impartiality than would a national court. Despite these advantages of arbitration, the development of contract law may suffer considerably by a withdrawal from the courts of litigation involving some of the most significant and difficult problems of the present day, all the more so because the reasoning in arbitral awards is usually not made public.

U.S. Uniform Commercial Code

Codification. Trade and commerce flow increasingly across national and state boundaries. In response to this there have been many efforts to unify the traditional legal systems. In the United States, the Uniform Commercial Code has replaced earlier uniform statutes such as the Sales Act and the Negotiable Instruments Law; by 1970 it had been adopted by every state including, although in part only, Louisiana. Internationally, the decade of the 1960s saw significant progress toward uniform regulation of the law of sales. The creation of a uniform body of substantive rules is, of course, easiest when the communities involved have roughly similar rules and principles. In addition, the greater the volume of multistate transactions, the greater the pressure for uniform regulation. It is understandably easier to achieve a Uniform Commercial Code within the United States than to create such a system internationally.

When a transaction has a significant relationship with more than one legal order, difficult problems of private international law often arise with respect to which law shall govern. A kind of halfway point between legal diversity and unification—the creation of uniform rules for choice of law—is of some help, and in this area the Hague Conference on Private International Law has done significant work. (A.T.v.M.)

Agency

The law of agency governs the legal relationship in which one person (the agent) deals with another (third party) on behalf of still another (the principal). The competent agent is legally capable of acting for this principal vis-à-vis the third party. Thus, the process of concluding a contract through an agent involves a twofold relationship. On the one hand, the law of agency is concerned with the external business relations of an economic unit and with the powers of the various representatives to affect the legal position of the principal. On the other hand, it rules the internal relationship between principal and agent as well, thereby imposing certain duties on the representative (diligence, accounting, good faith, etc.). The two relationships need not be in full conformity. Thus, an agent's effective powers in dealing with outsiders may extend to transactions that he is under a duty to his principal not to undertake, leading to a situation characterized as "apparent authority."

Agency is recognized in all modern legal systems as an indispensable part of the existing social order. It fulfills the most diverse functions in both public and private law; in particular, it assists in organizing the division of labour in the national and international economy by making it possible for a principal greatly to extend his individual sphere of activity by having one or more persons act for him. In addition to the individual principal, a principal may be composed of a group of persons carrying on a trade or business by way of a partnership, a registered company, or another kind of corporate entity. The need for legal representation in some form has therefore increased as business units have come to involve transactions conducted at a distance (through the use of factors, or commercial agents) or have grown in size (as in the case of the firm, the house, and the corporation). Continental law additionally allows the use of legal representatives, such as the father, mother, guardian, or curator (*curateur, tuteur*), to enable minors, insane persons, and other legally incapacitated persons to act. Although a similar category of "authority by law" is not unknown in common law, powers based on family relationships are scarce and appear in only a few cases.

HISTORICAL DEVELOPMENT

Roman law. Because concepts grow primarily out of specific situations that have occurred and social needs that have arisen, the doctrine of legal representation developed differently in different times and places, sometimes even within a single legal system. At first it seemed unthinkable that an agent, by making a contract with a third party, could create obligatory rights and duties between a third party and a principal. Even the official law of the Roman Empire never fully recognized the principle of representation. The explanation for this rejection lies mainly in the early Roman conception of a contractual obligation as a personal relationship binding the parties in some quasi-mystical way. This type of relationship allowed creditors in some instances to seize the possessions—and in very early times also the person—of the debtor. Usually the formation of such a relationship between two parties took place in a solemn ceremony at which both parties had to be present, certain formal words spoken, and definite acts performed. In such a situation it was impossible to confer right or duties on a third party. On the other hand, the head of the household could transact business through his slaves or his dependent sons, who were not conceived of as agents but as "long arm" extensions of the contracting master or father. Because of the wide prevalence of slavery, there was no great need for a true agency relationship. As Roman law later developed, the formalities connected with creating legal relationships became less important, and the need for personal representation in commerce increased. In the meantime, however, legal theory and practice had developed so many ways to evade the problem that there was no longer an urgent need for Roman law to overcome its stark conservatism and to develop a legal institution that it had earlier opposed.

Conception of contract

Medieval influence of canon law and Germanic law. Labouring under the influence of Roman law, legal development in the Middle Ages strove to overcome disadvantages in daily commercial life caused by the Roman rejection of the principle of agency. Through the efforts of legal scholars (glossators and commentators), Roman law was further developed by means of extensions, emphases, and exceptions—a process already sanctioned by the Romans themselves. Additional impetus for change came from Roman Catholic Church canon law. Although manifestly structured after Roman civil law, canon law had its own special development, influenced by Hebraic theological concepts. Certain writers succeeded as early as 1200 in constructing a type of agency relationship based on the position of *procurator,* a relationship intended to solve the representation problem in all except legal matters. The issue nevertheless remained in dispute.

About this time, the doctrine of principal and agent developed in England as an outgrowth or expansion of the doctrine of master and servant. Anglo-Norman law created the figures of *ballivus* and *attornatus.* His position in the household of his master empowered the *ballivus* to transact commercial business for his master, reminiscent of the power of the slave to bind his master under Roman law. Later the *ballivus* was given more authority, especially in his frequent role as land administrator, gradually becoming competent to act independently for his master. On the other hand, the *attornatus,* originally just a representative of one of the parties in litigation, soon assumed a position of broader importance. Certain contracts were effective only when made in a judicially prescribed manner. For this reason, the formation of this type of contract always had to be concluded in a court proceeding in which an *attornatus* represented each party. This was the beginning of the role of the *attornatus* as a general agent.

Modern developments. Recognition of the principle of agency in the field of civil law was finally achieved in continental Europe during the ascendancy of natural law in the 17th century. By this time, however, new objections grounded in state law, feudal law, and the question of the general reasonableness of agency had to be overcome. Hugo Grotius in his best-known work, *De Jure Belli ac Pacis* (1625; *On the Law of War and Peace*), explained that on the basis of his mandate a *procurator* could acquire rights directly for his principal. He thereby overcame the Roman rule that allowed slaves and dependent sons, but not free persons, with two exceptions, to act directly for the head of the household. Grotius simply maintained that this rule did not contradict natural law. In another work, *Defensio Fidei Catholicae,* Grotius added, in a theological context, that the principle of agency is based not on essential natural law but on nonessential natural law; that is, agency is not demanded by the nature of things but must only correspond to and be adapted to the nature of things.

The codifications of the subsequent age of rationalism recognized the principle of representation. The Prussian Civil Code (1794), the French Napoleonic, or Civil, Code (1804), and the Austrian General Civil Code (1811) nevertheless regarded agency as an aspect of mandate and the power to act as an agent to be derived solely from that concept.

The master–servant theory

A similar dualism governed the English development of the doctrine of principal and agent, in large part because of the new doctrine's dependence upon that of master and servant. The jurists of the 18th century, employing the phrase *qui facit per alium, est perinde ac si facit per se ipsum* ("whoever acts through another acts as if he were doing it himself"), made much use of the fiction of identity. In contrast, 18th-century commercial law, which at the end of the Anglo-Saxon period had split off from the main body of common law and had been allowed to develop under less-stringent controls, modified numerous feudal common-law concepts. In particular, it brought a legal flexibility allowing the law of agency, which had grown from three independent common-law roots—one relating to actions of debt and assumpsit (informal contract), another connected with the action of account, and the third deriving from deeds—to adapt itself to the peculiar problems and requirements of an individual case. Through this commercial influence, the concept of undisclosed agency, an especially important mode of representation in commercial transactions in which the agent appears to be the principal, developed alongside open agency representation in English law. Thus, even when a principal with the real economic interest in the transaction remained concealed, claims for and against him began to be recognized, whereas in continental law the necessity for the agent to act openly in the name of the principal retained a more fundamental importance in the general doctrine of contracts. English law attempted to relate agency rules more closely to the everyday needs of the principal–agent relationship, in contrast to the conceptually systematic and more narrowly confining treatment of agency in continental European law. As Alfred Conard, a modern writer, has observed, agency came to be a legal institution in English and North American law, "knitting together the whole subject of the employment of one man by another, whether for one job or for life, and whether on commission or on payroll."

While the first practical treatises on agency in England, by William Paley, and in the United States, by Samuel Livermore and Joseph Story, appeared early in the 19th century, in continental Europe the doctrines of the theoretical jurists continued to play their traditional leading role in the development of this branch of the law. A particularly important distinction in the European law of agency was made in the second half of the 19th century by the legal scholars Rudolf von Jhering and Paul Laband. Before them, agency was viewed solely in terms of the relationship binding the principal, the person being represented, and the agent, the person representing; that is, agency was equated with the relationship created by the mandate given to the agent. Jhering and, especially, Laband were the first to distinguish sharply between the agent's power to create legal rights and obligations for his principal and the inner contractual relationship governing the personal rights and duties between principal and agent. This distinction was largely adopted by the continental European legal systems and was codified in several countries.

Nineteenth-century legal theory also strengthened the so-called disclosure principle (*Offenheitsprinzip*), according to which the acts of an agent have direct legal implications for the principal only when the agent makes it known to the third party by acting in the name of his principal that he acts for him and not for himself. Only in such a case can there be direct representation—where the principal alone is the party to the contract—in contrast to indirect representation. Thus, this distinction became, to varying degrees, a common element in the individual civil-law systems.

Anglo-American law, more realistic in this area and standing on different dogmatic grounds, was not influenced to the same degree by such principles, although it was also long plagued by refined doctrinal distinctions, such as the contrast between general agents, with whom an outsider could deal with moderate safety, and special agents, whose powers the courts viewed very narrowly. At this stage of development the remaining feudal elements in the English law of agency, whose prototypes had remained the servant and the steward of a lord, were largely eliminated. Because of the increasing complexity of most modern commercial transactions, the significance of the various types of representatives grew to the point where agency had become one of the most important of modern legal relationships. The agent had come to be seen as an instrument of the principal without personal responsibility, a view which developed still further the responsibility of the principal for the misconduct of his agent.

Reflecting the generally more realistic attitude of contemporary private law, the modern treatment of agency has retreated from the somewhat artificial conceptualism of the 19th century and has emphasized the actual commercial and social context. Tendencies both to apply general solutions to broad types of cases and to provide specific solutions in the more unique individual cases have produced further refinements in the law of agency. The continuing shift to more complex business units, accompanied by developments in marketing practices, has required solutions more appropriate to modern business necessities. In addition to a reassessment of the relationship between mandate and authority to act for the principal, new questions are being asked concerning the proper role of this kind of authority in relation to the legal and social functions of representation that it serves. Accordingly, the mere fact of one's external position, which without authority cannot properly be used, has been increasingly acknowledged as an independent ground for recognizing an agent's power of representation.

Authority and action

At the same time that modern society's concern for third persons' rights has led to recognition of powers of position, thereby making an agent's power of representation less dependent upon the will of his principal (the latter being bound in these cases whether or not he wishes to be), a better view of private autonomy, as expressed in the act of authorization, sees a fundamental connection between the agent's act and the will of his principal in the legal transaction as a whole. In this view the authorization by the principal and the act of the agent are not two independently complete legal transactions carried out in isolation from each other. Rather, they are two parts of an extended legal event, which would consist of a single act if no agency were involved. Thus, when the principal uses the help of an agent, the offer, which is usually made by a single act, becomes divisible into two parts (the authorization and the action of the agent). Without such a connection, agency in private law, as it is understood today, would be largely incompatible with the principal's own autonomy.

THE RULES OF DIFFERENT SYSTEMS

In order to compare agency in continental and Anglo-American law, the principal types of agency that have developed in practice should be noted.

The continental "commercial agent" and his functions. The four main types of agent engaged in business activities in continental Europe can be identified as follows:

Commission agent (German *Kommissionär,* French *commissionaire,* Italian *commissionario*). The commission agent accepts or sells goods for the account of his principal, but in his own name. He is independent of his principal, has a claim for his commission, and, except in France, has the right when dealing with certain goods to conduct the transaction as he sees fit. The forwarding agent (German *Spediteur,* French *commissionaire de transport,* Italian *spedizioniere*) ships goods in his own name for the account of his principal and therefore is a type of commission agent.

Commercial agent (German *Handelsvertreter;* French *agent commercial,* or *voyageur,* or *représentant et placier;* Italian *agente*). The commercial agent negotiates and concludes contracts on behalf of his principal. Although the degree of his independence from the principal varies, he is never totally independent. While Italian law limits such an agent's activities to a specific geographic region, German law has such limitations only for the *Agent,* a subclassification of the general commercial agent who remains unrestricted in this regard. Apart from several protective rules, the French commercial agent is subject to the general rules governing the mandate.

Broker (German *Mäkler,* French *courtier,* Italian *mediatore*). The broker is a business agent who is completely independent of his principal. In the area of employment brokerage or placement services, most European countries have passed special regulatory legislation to protect the interests of those persons using such services to seek employment.

Sales representative. The sales representative is a dependent employee of a merchant who concludes contracts for the merchant outside the business establishment. Most European legal systems have no special provisions governing such an agent but rather treat the position under the general rules governing dependent commercial employees (*e.g.,* those governing German shop assistants, or *Handlungsgehilfen*).

The variety of Anglo-American agents. Various kinds of agency relationships are evident in Anglo-American commercial life. The factor and the broker are the most common mercantile agents dealing in transactions involving personal property. The factor is entrusted with possession of the chattels to be sold, or the documents of title thereto, and is empowered to conclude the sale at the best price obtainable. The broker, on the other hand, has no possession of the object of sale but is empowered to make contracts for the purchase or sale of personal property on behalf of his principal. More limited are the powers of the real estate agent, who may show the land and state the asking price to the potential buyer without ordinarily being empowered to make further representations. The store salesman is similarly restricted in his power to represent his principal and can usually do no more than make customary warranties and sell at the price fixed by his employer. In contrast, the traveling salesman not in possession of the goods normally has authority only to take orders, which are in effect mere offers to buy and are not binding on either party until the principal has accepted. The auctioneer is usually empowered to do no more than sell at the highest price bid.

Managerial and administrative agents

The common function of a second large class of agents is managerial or administrative. The manager of a business has the widest authority of all business agents and normally has complete control of all normal operations of the business. The agent employed to manage investments has a duty to deal only as would a prudent investor with reference to the principal's personal financial situation. The powers of the agent entrusted with his principal's land or chattels is more administrative; he has full power to protect the property and to make it profitable, but he may not sell or encumber it. Finally, the attorney at law may do what is necessary to advance his client's interests and may incur the necessary expenses in so doing but, unless specifically authorized, may not release any substantive rights of his client.

Apparent authority and related questions. If the principle of private autonomy were uncompromisingly applied to the law of agency, only an actually authorized agent could create legal rights and obligations for his principal. The will of the principal would then precisely define the boundaries of the agent's competence; however, a third party cannot always make a reliable determination of whether the agent has acted within the scope of this authority. Rather, he must often rely upon the principal's manifestation of the agent's authority, which may go beyond the more restricted authority actually communicated by the principal to the agent. Often he must rely upon the fact that the agent holds a certain position, such as wife, partner, or employee, in the belief that the normal incidents of authority implied by such a position are present, even though special restrictions on the agent's authority may have been made. And sometimes a third party must rely upon an earlier declared authorization that has since been revoked internally by the principal. Obviously, a third party cannot be expected to check all the details of the agent's internal authorization, especially since representation makes sense only when it functions efficiently and since use of an agent is also for the benefit of the principal. Therefore, the risk that the agent has acted without power of representation must be apportioned between the principal and the third party, with individual factors determining who must bear it in a particular case. This compromise is the subject of the doctrine of apparent authority.

It is obvious that the expression apparent authority is inappropriate since in none of the above cases is the competence of the agent illusory. With respect to the third party, at least, the authority of the agent is no less real than it would be if it were in conformity with the will of the principal. The distinction between apparent and real lies rather in the justification of the agent's acts in relation to the principal. In contrast to apparent authority, real authority is more than mere power of an agent, for along with the agent's legal power "looking out" (*posse*) stands his privilege "looking in" to the lawfulness of his conduct (*licere*). In the case of a real authority, power and privilege to exert the power are coextensive as against the principal, while, in the case of apparent authority, the agent has only the external legal power to perform without the corresponding internal justification vis-à-vis the principal.

Although the expression apparent authority spread from continental Europe, where it was used by, among others, the influential French classicist Robert Pothier (1699–1772), into English law, where Lord Ellenborough applied it in *Pickering* v. *Busk* (1812), its development in the two legal systems was independent and went systematically different ways. Nevertheless, the results obtained in concrete situations by the courts of both systems often seem rather similar. A similar disparity marked the development of the doctrine within Europe itself, with the formulation of different rules in the French system, which largely identifies authority and mandate, and in the German system, which distinguishes the two ideas. Another general tendency that has arisen in both legal systems has been to objectify the more subjective aspects of the doctrine, thus making the legal consequences more certain and predictable. Today both Anglo-American and continental courts permit the legal consequences of an agency relationship to result even when the principal is not willfully responsible for the appearance of authority.

Similar situations can also lead to opposite approaches in some matters. Here also it depends in the first instance upon whether the principal or the third party must bear the risk, since in general an agent acting without authority is not personally responsible. The question arises as to whether and to what degree the authority of the agent to bind his principal is affected by the death of the principal. According to the traditional English view (*Campanari* v. *Woodburn;* 1854), this event automatically terminates the agent's powers, irrespective of whether the agent or the third party knew or should have known of the death. While the rule originally might have rested upon the "fiction of identity" between principal and agent, today it is supported more by the technical argument that what a dead man cannot himself do, he cannot do through another, with further support garnered from the so-called doctrine of the meeting of the minds that is necessary for concluding a contract.

The death of the principal

In contrast, the continental legal systems have evolved less doctrinaire solutions to this question based more on considerations of protection of the concerned parties. Article 2008 of the French Civil Code even goes so far as to treat all transactions of an agent who acts in ignorance of the death of his principal as valid. The more balanced solution offered by the courts on the Continent, however, is to make the good faith of the third party the determinative factor, since in the usual case the real interest of the agent is simply to avoid any personal obligation.

Disclosed and undisclosed agency. Continental European laws restrict the application of agency rules to cases where the agent acts openly in another's name. Thus, French jurists infer from article 1984 of their Civil Code, according to which agency is the act of the agent *pour le mandant et en son nom* ("for and on behalf of the principal"), the negative conclusion that in case an agent does not disclose that he is acting as an agent for a principal, the consequences touch only the "agent" himself. The hidden principal is not concerned by the effects of the transaction at all. Section 164 of the West German Civil Code expressly provides that "an agent, who acts without disclosing the fact that he is acting as agent, is the only one to acquire any rights and is exclusively personally liable."

Undisclosed principal

In contrast to the continental view, when an agent contracts in his own name without disclosing his principal, the common law allows the undisclosed principal under certain conditions to sue or be sued by the third party. Such conditions include that the agent had power to make the contract and that the parties eventually learn their respective identities. This wider concept of agency has no counterpart in continental legal tradition.

The use of this basic doctrine in the common-law countries gives rise to questions regarding the identity of the undisclosed principal, the election of remedies that must be made by the third party, the extent of the respective liabilities, the right of the third party to setoff (the amount of its own damages from any sum that might be awarded it), etc. A solution to these conflicts of interests must in final analysis rest upon an evaluation of the extent to which the relationship between the undisclosed principal and the agent should influence the contract made by the agent with a third party.

The external, unilateral act of authorization. It is still a characteristic feature of the French Civil Code and of other codifications following its model (for instance, those of Spain, Portugal, Romania, and Brazil and other Latin-American countries) that agency is not recognized as an isolated institution. These legal systems conceive of agency only as a subordinate instance or external effect of mandate. The result is that they consider the power to act as an agent as a mere part of mandate and do not have a general concept of authorization as a distinct legal institution. In contrast to this approach, the more modern codifications of Scandinavia and of such countries as West Germany, Switzerland, Japan, Poland, Czechoslovakia, Italy, and Greece draw a sharp distinction between the unilateral organizational act on the part of the principal authorizing the agent to act and the internal contractual relations between the principal and the agent. This distinction, one of the major achievements of 19th-century European legal scholars, is also followed by modern English and American legal writers, even though the classical concept of mandate is unknown in the common law.

The insight that authority can exist independent of the underlying contract, and even without it, opens several new practical possibilities. For example, it explains the rule that the authorization of a minor can be valid if he is in fact mentally and physically capable of transacting business, even though the minor is not competent himself to conclude the (internal) contract of employment for himself that establishes fiduciary duties. Thus, the minor is not subject to the liabilities *ex contractu* (from or out of a contract) of an adult agent although he is authorized.

A few countries still prescribe a special form for every authorization. According to the Russian (RSFSR) and Brazilian codes, for example, an authorization must be given in written form. This general rule, which has been mitigated to some extent by the courts, is influenced by a distrust of informality and by a concern "to protect the individual and to be better able to examine the lawfulness of a legal transaction." The opposite tendency is followed to the extreme by the German Civil Code. It not only adopts the general principle of "freedom of form"—with the rare exception that a special form is required in those cases where a statutory article expressly prescribes it—but even states that the granting of authority need not be in the form prescribed for the judicial act to which the agency power relates. In spite of this express statutory language, West German courts do require compliance with formalities in certain extreme situations under pressure of practical necessity, in order to avoid frustration of the aim of the formal requirement in the principal contract. This attitude thereby approaches the English rule according to which, in most cases, no particular formalities are required, even if the agent is to make an agreement for the sale or lease of land, an agreement that must be in writing. More attention to the connection between the authorization and the act of the agent appears in provisions like that of the Greek Code of 1940, which says that the authorization must take the form required for the legal transaction for which the authority is issued unless the particular circumstances lead to a different solution. Other modern codes like those of Poland and Italy do not even contain the "unless" clause and prescribe only that an agent's power of attorney must be granted with the formalities prescribed for the contract that is to be entered into by the agent.

The internal agency contract. The internal bilateral relationship between principal and agent—apart from agency of necessity—rests upon what is commonly termed a "contract of agency." This term encompasses all sorts of contracts of employment including gratuitous services (in those cases in which the agent is willing to render a friendly favour).

This internal relationship between principal and agent must begin with an agreement, since otherwise the principal would have no rights against the agent—even the right to beneficial service presupposes the agent's consent to serve. The consent of the principal may be given expressly by a written power of attorney or implied by his conduct, such as an established course of dealing. Generally, the formation of the contract of agency requires no formal ritual. Moreover, in many cases the obligation to act as agent is only a part of the greater complex of duties that forms a person's job, profession, or office (for example, those of factors, bailiffs, attorneys, or brokers).

Consent of principal

The basic principle of agency is that the agent, in fulfilling his obligation, concludes legal transactions on the part of his principal. When these transactions result from the authorized legal acts of the agent, the result is that only the principal is bound by them. The competent agent himself remains outside the contractual relations with the third party (but compare "undisclosed agency" above) provided he has not also contracted for himself at the same time—for example, when the third party is not content with the principal's credit and insists upon having the agent's liability in addition to the principal's; in such a case the agent binds both his principal and himself. This general rule that the agent does not become involved may change in the exceptional case of an authorization conferred for the benefit of the agent. Thus, such an agent has been held personally liable by German courts for mistakes made in the course of contract negotiations (*culpa in contrahendo*).

Rights and duties between principal and agent. Continental European codifications generally do not treat the contract of agency as a special, separate type of contract. The standards governing the principal–agent relationship must therefore be derived from the general legal rules governing the mandate, the contracts for performing work, employment contracts, and partnership contracts, together with the more specific rules (*e.g.,* of the commercial codes) regulating particular agency relations and the individual transaction at hand.

Differing from this general approach, the common law has recognized a number of more specific rights and duties between principal and agent. The principal must provide the agent with a regular opportunity for service under the contract of employment and has a positive duty to aid, and not inhibit, the agent's performance of such service. He must compensate the agent for his services, provided they are not gratuitously rendered (the agent's right of remuneration), and must account for those amounts due to the agent, including indemnification for economic loss suffered by the agent on account of the agency relationship (right of indemnity and right of lien by way of charge on the principal's goods in his possession). The principal also has the more abstract duty to conduct himself so as not to harm the agent's reputation.

Remuneration, indemnity, and lien: the agent's rights

The primary duties of the agent to the principal are those of care, obedience, and loyalty—similar to those of a trustee. More specifically, the agent must act solely for the interests of his principal and therefore must account for any financial benefit ("secret profit") he might derive from a transaction. It follows that it is a violation of an agent's fiduciary duty to his principal to have unrevealed interests adverse to those of his principal. He must not compete with his principal and may not use or disclose confidential information except for his principal's benefit. If an agent has received money or other property from or for his principal, he must account for it. An agent also may not normally delegate his task to a subagent, since the principal is assumed to have placed his confidence in the person of the agent and not in a subagent. Finally, the agent also has the abstract duty of conducting himself so as not to bring disrepute upon the principal.

Liability of principal for acts of agents. Continental European law classifies the undertaking of transactions in the place of another as agency only when the transactions are legal. It excludes other acts, including unlawful acts, so that when dealing with the law of agency, the rules concerning the liability of a master for the torts of his servant do not come into consideration.

Following the lead of Justice Oliver Wendell Holmes, Jr., the opposite view has been taken in English and American literature. The predominant opinion treats the liability of a master for the tortious conduct of his servant as a part of agency law since these cases logically come within the maxim *qui facit per alium, facit per se* ("he who acts through another, acts himself"). The doctrine of *respondeat superior* ("that the master must answer") is therefore treated as a part of agency law, even though the rationale behind the master's liability is that he has assigned to a servant under his control a task that unforeseeably results in damage to a third person. On the other hand, consequences flowing from an agency relationship rest on the idea that an authorized person performs legal acts within his competence not on his own behalf, but for the principal. A growing view in the modern literature recognizes, however, that "the two basic relations are strange bedfellows."

According to English and American law, the liability of a principal for his agent's torts in the ordinary course of his employment depends upon the existence of a master–servant relationship. If the principal and agent stand in a master–servant relationship to each other, the master is vicariously liable for his servant's tortious conduct committed within the course of his employment irrespective of any personal fault on his own part. This doctrine of *respondeat superior* arose from the belief that since the head of the household or economic enterprise exercises control, he should pay for the harm caused by its members.

Such a relationship presupposes a degree of physical control by the master over his servant, which may exist, for example, over the driver of the principal's delivery truck but is lacking over the principal's factor or broker. The other major limitation to the doctrine is that the servant's tortious conduct must be committed within the servant's course of employment. Generally, this includes conduct that is not a serious departure from that authorized, both in manner and space, by the master and that is actuated at least in part by a motive to serve the master. Under this principle, the master's vicarious liability extends in some instances even to claims arising from the servant's intentional torts.

If no master–servant relationship exists between the parties, the principal's liability for the torts of his agent is far more restricted. For example, a principal is not liable for the harm caused by the negligent physical conduct of his nonservant agent. He is liable, however, for any other torts of his agent that either are authorized or the commission of which normally are made possible by the agent's position. The chief example of such conduct is tortious misrepresentation by the agent.

Termination of agency. With the termination of actual authority, questions often arise as to whether the authority was usual, was special or general, arose out of necessity or was also apparent. Questions concerning the external and internal agency relationships create special problems, particularly in this area. A principal's termination of the agency relationship does not necessarily have the same effect vis-à-vis the agent as it has with respect to the third party. It may be that the principal can terminate the real authority but not the apparent authority toward a third party.

Termination in breach of agreement

In theory, the principal may revoke the real authority of his agent or the agent may terminate his authority by renunciation or revocation at any time. If the revoking party has previously agreed not to do so, the revocation remains valid, but the other party may maintain a breach of contract action against him. The agency relation can also terminate by operation of law in case of death, insanity, bankruptcy, or war, particularly where the principal is concerned. The question then arises as to whether the unknowing agent is protected by the law, since he might be personally liable on the contract.

A major exception to the principal's power of revocation occurs when the agent possesses "authority coupled with an interest." Here the agent who has part ownership in something that is to be disposed of has power also to dispose of his principal's remaining interest in this thing. Such a power for the purpose of effectuating any security or of protecting or securing any interest of the agent cannot be revoked without the agent's consent. (W.M.-F.)

Commercial transactions

Law of exchange of values

The law of commercial transactions comprises the core of the legal rules governing business dealings. The most common types of commercial transactions, involving such specialized areas of the law and legal instruments as sale of goods and documents of title, are discussed below. Despite variations of detail, all commercial transactions have one thing in common: they serve to transmit economic values such as materials, products, and services from those who want to exchange them for another value, usually money, to those who need them and are willing to pay a countervalue. It is the purpose of the relevant legal rules to regulate this exchange of values, to spell out the rights and obligations of each party, and to offer remedies if one of the parties breaches its obligations or cannot perform them for some reason.

The law of commercial transactions thus covers a wide range of business activities. It does not, however, govern such essentially noncommercial relationships as those involved in succession and family law. Historically, land was of such prime importance that it was not subject to frequent disposition and therefore was also excluded from the category of commercial transactions.

In some countries the term commercial transactions is merely descriptive. In Anglo-American law especially, it is merely a collective name for those rules that relate to business dealings. The term itself has no legal consequences. It serves only as a convenient and illustrative shelter under which certain legal rules may be assembled.

Many countries, however, have established a technical concept of commercial transactions with precise definitions and important legal consequences. This is most often the case in the civil-law countries. In these countries the term commercial transactions thus has more than a merely descriptive function. It designates in part those rules that are peculiar to commercial transactions. In France, for example, bankruptcy is open only to individuals who are merchants and to business organizations, and there are special rules applying to commercial cases. In West Germany, similarly, the general rules on consumer sales are in part superseded by special rules on commercial sales. A commercial transaction thus results in a number of specific legal consequences that differ from those of ordinary consumer transactions. Such a special commercial regime exists usually because it is thought that the ordinary citizen should not be exposed to the rigours of commercial rules that presuppose a knowledgeable, versatile individual who does not need as much protection against the legal risks and consequences of his dealings.

In those countries in which specific legal consequences attach to commercial transactions, it is necessary to develop a precise definition of what constitutes a commercial transaction. Although such definitions are more or less closely related, they are peculiar to each country. Most of them, found usually at the beginning of a special "commercial code," combine two elements: definitions of a "merchant" and of a "commercial transaction." In certain countries—West Germany, for example—the emphasis is on the definition of the merchant; in others, such as France, the emphasis is on that of the commercial transaction (*acte de commerce*). This latter criterion, the so-called objective test, was adopted in the 19th century for ideological reasons, the French wanting to avoid any repetition of the pre-Revolutionary differentiation of legal rules according to the social condition of persons. But whatever the test, the results are similar, for the gist of the various definitions is that a transaction is "commercial" if it is concluded by a merchant in the exercise of his profession.

Socialist economic law

It should be mentioned that the Socialist countries of eastern Europe and of East Asia do not recognize the concept of commercial transactions. This is because such a concept is somewhat suggestive of a capitalistic form of the economy. The implication would be the more invidious since the state, directly or through the various forms of state enterprises, is practically the only trader in Socialist countries. All the Socialist countries have developed a mixed system, combining more or less obligatory state planning with the autonomous fixing of the remaining terms of their exchanges. The breach of these so-called economic contracts may entail the payment of damages to the other contracting state enterprise, thus drawing the attention of the state bank to violations of plan and contract discipline. These special rules for dealings between state enterprises within each country and also for enterprises from different Socialist countries are called "economic" laws.

HISTORICAL DEVELOPMENT

Only a few traces of rules on commercial transactions in antiquity have survived. The most notable is a rule developed by the seafaring Phoenicians and named after the island of Rhodes in the eastern Mediterranean. The "Rhodian Law" provided that losses incurred by a sea captain as a result of trying to save ship and cargo from peril must be shared proportionately by all owners of cargo and by the shipowner. If, for example, one merchant's cargo was thrown overboard in order to save the ship from sinking, the loss would be shared among the shipowner and all the other merchants with cargoes aboard. This rule applied in the entire Mediterranean and is today known in the maritime law of all nations as general average.

Another important rule, also of maritime character, arose in connection with the maritime loan that developed in Athens. A capitalist would lend money for a marine trading expedition. The loan would be secured by ship and cargo, but repayment of the capital and payment of interest were conditional on the ship's safe return. The interest rate of 24–36 percent, considerably beyond normal rates, reflected the highly speculative risks involved. This transaction later developed into marine insurance.

Roman law influences

Much more is known of the commercial law of the Romans. It was in Rome that for the first time a separation developed between the ordinary civil law and special rules for foreign (that is, primarily trade) relations. Since the civil law applied only to Roman citizens, trade and other relations with and among noncitizens were subject to a separate set of rules—the jus gentium, or law of nations. The latter exhibited two traits that have become characteristic of the law of commercial transactions: it was more liberal than the strict rules of the civil law, and it was applied uniformly in various parts of the world.

As far as specific rules are concerned, the Romans received and preserved the two institutions of the general average and the maritime loan that had been developed earlier. They added two other rules of maritime law: the liability of the shipowner for contracts concluded by the ship's master (an early recognition of an agency relationship that was later generalized) and the liability of the ship's master for damage to or loss of the passengers' luggage and equipment. Innkeepers were charged with the same liability. Banking transactions and bookkeeping were well developed, and some prohibitory rules were enacted against capitalist excesses. Thus, the legal interest rate was lowered. In the postclassical period an attempt at achieving a "just price" was made by introducing a rule that a sale could be annulled by the seller if the price paid to him was less than 50 percent of the value of the goods sold.

In the Middle Ages the Christian Church attempted to enforce certain moral commands adverse to commercial transactions. The taking of interest for loans of money was considered income without true work and, therefore, sinful and prohibited. There was also an attempt to generalize the idea of a just price. Although both rules, and especially the former, influenced the law and the economy for centuries, neither of them finally prevailed in the secular world.

Emergence of the law merchant

Another feature of the medieval period was the development of a separate commercial law—the law merchant. Like the jus gentium of early Roman days, the law merchant was different from the existing ordinary rules that varied from place to place. The need for certainty and uniformity in the provisions governing trade motivated the growth of one set of rules for commercial transactions, valid everywhere in Europe. These rules were disseminated and applied in special courts conducted at the numerous international fairs held in various countries of Europe and attended by local and foreign merchants. The main sources of the law merchant were the customs of the most developed commercial communities of the time—the northern Italian cities. Later, in the 13th and 14th centuries, Italian, French, and Spanish cities made the first attempts at codifying certain branches of commercial law.

The medieval period saw the development of company and banking law. The *compagnia* and the *comenda,* forerunners of the partnership and limited partnership, were in frequent use. The Italians created a sophisticated system of bills of exchange used partly for the transfer and exchange of money, partly (by means of endorsement) for payment, and partly (by discounting) for credit purposes. They also invented bankruptcy as a method for dealing equally with an insolvent merchant's creditors.

In the period following the medieval era, but before the French Revolution, the law of commercial transactions lost its universal character. The birth of pronouncedly national states in Europe provoked a "nationalization" of the law. In 1673 and 1681 the French king Louis XIV enacted ordinances on land and maritime commerce. These were precursors of the French Commercial Code of 1807, which set the pattern for national codification of the law of commercial transactions in the Latin countries of Europe and America. In England the chief justice Lord Mansfield began from about 1756 to blend the law merchant into the common law. Only maritime law, although nationally codified, preserved some of its universal traits.

Of great consequence for the later development of commercial law was the foundation of colonial companies, usually through royal charter, for the exploitation and administration of the colonies of the European countries. The first, the Dutch East India Company, was chartered in 1602. Only such companies were able to attract the immense amounts of capital that were needed. The liability of each member was limited to his contribution, which was represented by share certificates that were transferable. Limited liability of shareholders and negotiability of shares were in fact fundamental to the operation of these companies. They were adopted and refined later into the most important vehicle of modern capitalism—the corporation.

ELEMENTS OF THE LAW OF COMMERCIAL TRANSACTIONS

In the 20th century, domestic as well as international commerce has experienced an expansion far beyond any earlier dimensions. With the multiplication of commercial transactions the demand for legal certainty has increased, especially for transactions across national boundaries.

Contractual relations. The first response to the multitude of practically identical transactions was the standardization of contracts. Printed standard contracts or forms lay down those provisions that are essential in the eyes of the drafting party. It depends upon the relative economic strength of the other party whether departures from the printed form can be negotiated. Trade associations as well as individual enterprises have developed and elaborated forms and standard contracts for their members.

The same technique of standardization has been adopted for international transactions. The forms and standard contracts of certain well-known trade associations, especially British ones, such as the London Corn Trade Association, are used by exporters and importers in many countries. The same is true of many shipping transactions. Even international bodies, such as the United Nations Economic Commission for Europe, have elaborated printed forms for certain international contracts. Apart from standardizing the contract practices of a particular party, these uniform conditions also help to bridge the gap between the many different national rules. They are a means of achieving partial uniformity of law for international trade.

International unification of commercial law

The development of uniform legislative rules for international transactions has been another distinctive feature in the 20th century. This trend resulted from the uncertainties to which international commercial transactions that came under two or more national jurisdictions were exposed. International conventions have resulted in the unification of numerous rules, especially in the areas of transportation, industrial property (patents and trademarks), copyright, and commercial paper (bills of exchange and checks). Less successful so far have been attempts in the fields of sale of goods and the conclusion of contracts.

Despite considerable progress in the field of unification, none of the uniform rules is really worldwide in scope, many being limited to a continent or to narrower regional groups such as the countries of the European Economic Community or the Council for Mutual Economic Assistance (of eastern Europe).

Sale of goods. The sale is the most common commercial transaction. All the rights that the seller has in a specific object are transferred to the buyer in return for the latter's paying the purchase price to the seller. The objects that may thus be transferred may be movable or immovable and tangible or intangible. (Patents are an example of intangibles.)

The essence of a sales transaction

Not all transfers of goods to another person for any purpose whatsoever constitute a sale. Goods may be transferred for use only (lease), for safekeeping or storage (bailment), as a present (gift), or in exchange for another good (barter). They may also be transferred as security. A sale is involved only if the seller intends to part with the object completely and conceivably forever and to receive instead a sum of money as the price.

The seller's complete parting with all his rights in the object sold means, in legal parlance, transfer of ownership to the buyer. One may say that the transfer of ownership for a price is the essence of a sale. But this rule applies only in countries with a free economy, not for the Socialist states in eastern Europe and East Asia, where the sale of goods among state enterprises transfers not ownership but merely the right of administration. Such sales are therefore designated as "delivery contracts." For most practical purposes, however, a sales contract and a delivery contract can be equated.

Obligations of the seller. The seller's duties are three: he must deliver the goods, transfer ownership in them, and warrant their conformity to the specifications of the contract.

Delivery of the goods sold to the buyer must be at the time and place and in the manner agreed upon by the parties. Nondelivery is sanctioned by the various legal systems in three different ways. Anglo-American law does not, in general, permit the buyer to sue for delivery of the merchandise but requires him to buy elsewhere and to demand damages from the original seller. The buyer is entitled to a decree for delivery (specific performance) only if damages are an inadequate remedy because the buyer cannot obtain substitute goods in the market. On the European continent, by contrast, a buyer may always demand delivery. Merchants do not usually go to the trouble of suing for delivery, however, but act voluntarily as their English and American counterparts are by law enjoined to act: they buy the same or similar goods in the market and then sue the nonperforming seller for damages. The measure of damages is usually the difference (if any) between the original contract price and the market price at the time of the substituted purchase. This covers the loss arising directly from the seller's nondelivery. Additional loss, such as expense arising from the substituted purchase or a loss on the intended resale of the goods, may also usually be claimed as damages from the seller.

Delayed delivery

From the point of view of the buyer, delayed delivery is connected with nondelivery in two ways. After the time for delivery has passed, the buyer may not know whether the seller is failing to deliver at all or whether delivery has merely been delayed. Further, the delay in delivery may be as harmful to the buyer's interests as outright nondelivery. This latter situation is particularly likely to arise if the agreed time for delivery was of the essence of the contract (that is, if it constituted so vital a stipulation that without compliance the contract could not be fulfilled). Even if the parties did not agree expressly that prompt delivery was crucial, such a condition may have been implicit because of the nature of the goods sold (for example, in a contract for the sale of raw materials subject to marked fluctuations in market price or for the sale of perishable or seasonal goods).

Countries differ considerably in the treatment of delayed delivery. Most legal systems require a more or less formal request for delivery or information by the buyer from the seller if a precise delivery date had not been agreed upon. If a precise time had been fixed but was not essential, such a request for information is usually unnecessary, except in France and some other Latin countries. But even if the buyer is not obligated to make inquiries of the seller, additional steps may be necessary in order to obtain remedies for nondelivery. In France and some other Latin countries, the buyer must bring suit for dissolution of the contract, and the judge may grant days of grace to the seller for performance. In West Germany, the buyer must grant the seller a reasonable period of time and declare unambiguously that he will refuse acceptance thereafter. Neither Anglo-American nor Scandinavian law protects the seller with such a period of grace. If the time element was crucial, the buyer's remedies in these countries are the same as for nondelivery. If, however, the seller did in fact deliver, although belatedly, the buyer may claim general damages for the loss arising from the delay.

Under certain circumstances the seller may be excused from his obligation to deliver on time. This is generally the case if prompt delivery becomes impracticable because of an unforeseeable and unavoidable obstacle. But if the seller owes a quantity of a certain kind of product and has not by the time delivery is due appropriated specific pieces for the purpose of delivery, he is rarely excused. In major contracts the parties usually make specific provisions concerning the conditions under which the seller is to be exempted from liability for late delivery.

Delivery must be accompanied by transfer of ownership to enable the buyer to enjoy full legal rights over the objects sold. The method of transferring ownership varies in two main ways. In most countries, ownership in a specific object is transferred with the conclusion of the contract of sale unless the parties agree otherwise. Such a transaction in Anglo-American law is called a "sale," as distinct from a mere "contract to sell." In the case of generic goods (any goods within a class rather than specifically designated goods; for example, 10 tons of coal), ownership cannot pass to the buyer until the seller has specified those goods which he intends to deliver (by transferring 10 tons of coal to a carrier for transportation to the buyer). But the parties may delay the transfer of ownership, perhaps until delivery to the buyer or until payment of the purchase price. If nothing has been agreed, the seller, although no longer the owner, may refuse to deliver or stop the goods en route to the buyer if the latter's solvency has become doubtful after conclusion of the contract. If the seller resells the same goods to a second buyer, the first buyer's claim to the goods prevails unless the second buyer has received the goods. In the central and eastern European countries ownership does not usually pass to the buyer until he receives the goods.

Although it would appear to be logical that a buyer cannot become the full owner unless the seller had unrestricted ownership, the demands of commercial expediency have carved out important exceptions in favour of a purchaser in good faith. Details vary considerably from country to country. At least between merchants, the acquisition of goods from one in possession of them who can in good conscience be regarded by the other as their owner, or at least as being entitled to their disposition, usually confers ownership on the buyer, even if the seller was not in fact the owner.

Remedies for defective title

The sanctions available to the buyer who does not obtain unrestricted ownership vary from country to country. Some countries impose upon the seller the outright obligation to procure ownership in the goods sold to the buyer. A violation of this duty is a breach of contract and opens the same remedies as those for nondelivery, including a suit for transfer of ownership. But in most countries the seller's obligation is limited to warranting "quiet possession"; that is, guaranteeing enjoyment of the goods undisturbed by claims of third parties. In some countries the warranty of quiet possession entitles the buyer who is sued by a third party to call the seller into the proceedings or even to turn the proceedings over to the seller so that the latter may defend the action. Everywhere the buyer may claim damages from the seller, covering not only the difference between the contract and the market price of the goods but also the expenses of defense against the claims of the third party. The buyer's rights are usually excluded if he knew of the seller's defective title at the time of contracting or if he became aware of it at some later time but nevertheless accepted the goods.

Caveat emptor

Goods sold must conform to the specifications of the contract as to their physical qualities, kind, and quantity. The rules on the delivery of goods of defective quality have a long history. Roman as well as English law originally denied the buyer the right of any claims as to quality under the doctrine of caveat emptor ("let the buyer beware"). This general rule did not apply, however, if the buyer had received express guarantees from the seller. Gradually the law developed various "implied warranties," the breach of which gave rise to certain special rights. As a result, the quality of goods is generally considered defective if they are unfit for the ordinary purposes for which such goods are used or unfit for the buyer's special purpose, provided the latter was known to the seller. As soon as possible after delivery the buyer must examine the goods for defects and must notify the seller if any are found. The buyer may then accept the goods but make a deduction from the purchase price for the defect. In most legal systems the buyer may alternatively reject the goods and dissolve the contract of sale. The buyer may also claim damages from the seller, but usually only under special conditions. A third remedy open to the buyer is to demand delivery of conforming goods, but this right is usually limited to generic goods. The buyer's rights are vitiated if he knew of the defect at the time of contracting or if he failed to avail himself of his rights immediately on delivery or within a limited time thereafter. Remedies for defective goods are often widely modified by contractual agreement between the parties.

Obligations of the buyer. The buyer's main duties are simple: payment of the purchase price and acceptance of delivery. Contemporary legal systems are no longer concerned with enforcing a just price. Only a few European countries (including Italy and France) still have rules on exorbitant prices and only in certain special fields. The buyer is strictly responsible for payment of the agreed price and cannot excuse himself by invoking his financial straits. Only war, revolution, exchange restrictions, and other unforeseeable and unavoidable obstacles to performance may, under certain circumstances, excuse the buyer from his duty to pay.

Just as the buyer is often unable to secure specific performance of the seller's duty to deliver, so the seller is not always able to enforce his claim for acceptance of delivery against a buyer who refuses to take delivery. Most countries do not object to such a claim, but in England and the United States the remedy is rather to refer the seller to the market: as long as he is still the owner of the goods, he should at least attempt to resell them at a reasonable price. Only if this is impossible or impracticable may he sue the buyer. In many other countries the seller, though not obliged, is at least entitled to resell the goods. The proceeds of the resale diminish the seller's loss; but the original buyer remains responsible for the difference. The seller may also, without actual resale of the goods, claim this difference as damages. If the buyer merely delays payment, the seller may usually claim compensation for any resulting loss. Very often this loss is calculated in a lump sum and takes the form of interest on the outstanding purchase price, the rate of which is in many countries provided for by statute. Additional damages for any further loss usually may be claimed. The buyer is, in general, excused from the payment of interest as well as additional damages if the delay of payment was due to unforeseeable and unavoidable obstacles.

The buyer's obligation to take delivery of the goods depends, as regards details, on the precise agreement of the parties: if steel plates, for example, have been sold "free on board vessel," the seller must load the plates on board the vessel named by the buyer. If the latter does not name a ship, the seller cannot perform his duty of delivery.

If the buyer fails to make provision for taking delivery, the seller still must preserve the goods, although he is no longer fully responsible for their fate. In many countries the seller may deposit the goods; in others he has the right to resell or a choice between the two. The proceeds of the resale take the place of the goods and have, therefore, to be paid to the buyer. The seller may claim damages arising from the buyer's breach of duty.

Mutual obligations of the seller and buyer. The duties of seller and buyer do not exist separately and independently from each other but are mutual and concurrent. Both parties assume duties in anticipation of the performance promised by the other party. It is a major consequence of the principle of mutuality of obligations that the duties of seller and buyer must be performed in general at the same time unless the parties agree otherwise. In international sales transactions it is often agreed that the seller must ship the goods to the buyer, so that the latter need not pay until he has received the goods and has thus been able to inspect them. Sellers may reestablish the time balance by demanding "payment against documents"; that is, payment when the buyer receives the documents of title, although the goods themselves may still be with the seller or in transit. The law everywhere protects the time sequence agreed upon by the parties by allowing a party to refuse its own performance as long as the agreed advance performance has not been made by the other party. The technical legal means used to achieve this result vary considerably. In exceptional circumstances the party that is obligated to perform first may refuse to do so. This may be justified if the other party's financial situation after conclusion of the contract has become so aggravated that payment is doubtful.

Time sequence of performance

Various countries differ widely in determining when risk for lost or damaged goods passes to the buyer. In several countries risk passes at the conclusion of the contract to sell; in others, notably France and England, risk is tied to the transfer of ownership in the goods; in West Germany and most eastern European countries, risk passes at the time of delivery; and in the United States and the Scandinavian countries, risk passes when the seller has essentially performed his duties. In all countries, the parties may, expressly or implicitly, agree to some other suitable arrangement.

Unification of sales law. The many differences in the sales laws throughout the world are a serious obstacle to an effective and smooth international trade. In view of the great volume of international trade, attempts at unification of sales law have been undertaken for many years. The most thorough results may be expected from a unification of the diverging rules on sales themselves. A more modest approach, however, has been to develop common rules on how to proceed when a conflict between the divergent national sales laws occurs. Efforts at unification have in fact followed both lines.

A considerable degree of unification of sales rules has been achieved by the wide acceptance of certain form contracts. But however successful some of these form contracts have proved, they have two important drawbacks: their validity depends on their acceptance by both contracting parties, and they cannot override the mandatory rules of national law.

These drawbacks can only be overcome by unifying national legislation. This method has been used with great success by the Socialist countries of eastern Europe. These have agreed, within the framework of the Council for Mutual Economic Assistance, on Uniform Conditions for Contracts of Delivery Between Foreign Trade Enterprises (1958, revised 1968 and 1975). The elaborate "conditions" have the force of law; the enterprises may not deviate from them except under special circumstances. For cases not expressly covered by the conditions, there is a uniform conflicts rule that declares the law of the seller's country applicable.

Other countries have had much less success. After almost 40 years of preparation, an international conference at The Hague adopted in 1964 a Uniform Law on International Sales. Under the auspices of the United Nations Commission on International Trade Law (UNCITRAL) a revised Convention on Contracts for the International Sale of Goods was signed in 1980 and entered into force on Jan. 1, 1988, in some countries.

A much more modest approach to the harmonization of legal divergences is the unification of the conflicts rules relating to international sales. A convention to this effect concluded in The Hague in 1955 has been ratified by eight European countries. According to this convention, the parties are free to choose the applicable law; if they do not do this, the law at the seller's place of business will, in general, govern the sales contract. The effect of these rules is, however, limited. They merely ensure that the courts in the participating countries will apply the same law to an international sales contract: the divergences between the different sales laws are not overcome.

Negotiable instruments. The negotiable instrument, which is essentially a document embodying a right to the payment of money and which may be transferred from person to person, developed historically from efforts to make credit instruments transferable; that is, documents proving that somebody was in their debt were used by creditors to meet their own liabilities. Thus a promise of A to pay B a certain sum at a specified date in the future could be used by B to pay a debt to C. This "negotiability of credit" was facilitated by the development of a variety of negotiable instruments including promissory notes, checks, and drafts (bills of exchange). These are in fact the most common negotiable instruments in use, and the following discussion will be confined to them.

Negotiable instruments are used for purposes of payment or credit and as security. Sometimes one instrument may perform all three functions. A typical "trade bill" used in connection with an inland or an export sale serves as an example of this: the seller, according to a clause of the contract of sale, may draw a bill on the buyer (that is, prepare a "promise to pay" that the buyer must sign) or, in the case of an overseas buyer, on a bank acting for the buyer, payment to be made within the agreed time (such as 30, 60, or 90 days after delivery). The buyer or his bank signs the bill as drawee and thereby becomes acceptor. On return of the instrument the seller may use this accepted bill to pay his own debts or may sell it to his bank (discounting). The buyer may also, although this is not typical for commercial transactions, draw a check on his own bank and send it to the seller.

Drafts as negotiable instruments

The most common and most complex form of negotiable instrument is the draft, or bill of exchange. It has been defined in England as an unconditional order in writing addressed by one person to another, signed by the person giving it (the drawer), and requiring the person to whom it is addressed (the drawee) to pay on demand or at a fixed or determinable future time a certain sum of money to, or to the order of, a specified person (the payee) or to the bearer. In the United States the definition is the same, except that an instrument may only be made payable "to order or to bearer." If the drawee assents to the order and accepts the bill, which is done by signing his name, or his name with the word "accepted," across the face of the paper, he is called an acceptor. The person to whom a bill is transferred by endorsement is called the endorsee. Any person in possession of a bill, whether as payee, endorsee, or bearer, is termed a holder and, if he is a bona fide purchaser, a holder in due course.

The basic rule applying to drafts is that any signature appearing on a draft obligates the signer to pay the amount drawn. It is the characteristic feature of a draft that it is not limited to the three-cornered relationship among drawer, drawee, and the named or unnamed creditor. Rather, the creditor may transfer it (for purposes of payment or borrowing) to a fourth party, and the latter may transfer it to a fifth, and so on, in a long chain. The means of accomplishing a transfer from one creditor to another is by endorsement or delivery. If an instrument is payable "to order," the signature (endorsement) of the transferor is required. The draft is then delivered to the new creditor. If the instrument is payable "to bearer," delivery alone suffices. Endorsement transfers the rights of the endorser to the new holder and also creates a liability of the endorser for payment of the amount of the draft if the drawee does not meet payment when the draft becomes due.

A failure to pay a draft must be more or less formally ascertained (in continental Europe through a formal "certificate of dishonour"). Upon due notice of dishonour, the holder of the draft may claim payment from any endorser whose signature appears on the instrument, and he in turn may claim from prior endorsers, from the drawee, and from the drawer.

The necessity of unifying the legal rules relating to negotiable instruments used in international trade has long been felt, and considerable success in this direction has been achieved. The principal rules in English law are laid down in the Bills of Exchange Act of 1882. This act spread through the whole Commonwealth and also influenced the United States Negotiable Instruments Act proposed in 1896 and eventually adopted throughout the United States. This latter act has since been replaced by article 3 of the Uniform Commercial Code. On the Continent uniformity between the French and the German approach was first achieved at two conferences held at The Hague in 1910 and 1912 and finally by two Geneva conventions of 1930 and 1931 on uniform laws for drafts, promissory notes, and checks. These latter agreements included some uniform provisions on conflicts of law. These have been adopted by most European countries and by many states in other parts of the world. Neither England nor the United States accepted these conventions, however, partly for fear of upsetting the uniformity already achieved in the Anglo-American orbit.

Documents of title. Whereas negotiable instruments embody a claim for the payment of money, documents of title embody claims to goods. The most common such documents are the bill of lading and the warehouse receipt.

Bills of lading

A bill of lading is a receipt for goods delivered for transportation by a ship. On receiving the goods alongside or on board, a dock or mate's receipt is issued and is later turned in for the bill of lading proper. The bill of lading may certify receipt of the goods either on board the ship ("shipped on board") or alongside ("received for shipment"). This latter form of bill of lading is less valuable since it does not prove the fact and date of loading. Apart from proving receipt of the goods to be shipped, the bill of lading incorporates the terms of the contract concluded between the carrier and the consignor for the transportation of the goods to the port of destination. A great many of the printed clauses on a bill of lading purport to excuse the carrier from liability for delayed delivery or from liability for damage to or loss of the goods. These clauses are valid, however, only if and insofar as they comply with the applicable national law or, in the case of ocean transport, with the Brussels Convention on Limitation of Liability (1923, amended 1968) that incorporates the Hague Rules, which have been adopted by the major shipping nations. Subject to these contractual terms, the consignee (the person to whom the goods are being shipped) may, by virtue of the bill of lading, demand delivery of the transported goods at the port of destination. In the simplest case the consignor sends the bill of lading by airmail to the consignee so that the latter may claim the goods on the arrival of the ship. The carrier may only deliver the goods to a person holding a duly negotiated bill of lading.

A bill of lading and the claim it represents may be transferred to another person by endorsement and delivery of the document. If made out to bearer (which happens rarely), the bill may even be transferred by mere delivery. By such transfer all the rights and obligations embodied in the document are transferred to the new holder. The latter is entitled to demand delivery of the goods unless the carrier proves that the holder knew or through gross negligence was unaware of the transferor's lack of title to the bill. In contrast with the rules on negotiable instruments, an endorsement of a bill of lading does not make the endorser liable for any default of the carrier or previous endorsers. The bill represents the goods, and transfer of the bill is, therefore, equivalent to delivery of the goods to the transferee.

It depends on the intention of the parties whether ownership in the goods or merely a security interest in them is to be transferred. A security interest is typically acquired by a bank, which gives credit on the security of the shipped goods. The above rules on bills of lading, though not formally unified, are essentially the same in all the seafaring nations. Most of them apply also to bills of lading issued in river navigation.

Warehouse receipts

The warehouse receipt is a document that shares the essential traits of a bill of lading, except that the duty to transport the goods is replaced by an obligation to store them. This receipt also embodies the claim for delivery of the goods and may, therefore, if made out to order, be transferred by endorsement and delivery. According to the intention of the parties, such a transfer may pass ownership in the stored goods or create other rights, such as a security interest, in them.

Letters of credit. Of great importance in international trade is the letter of credit. A letter of credit is essentially an authorization made by a buyer to his agent (usually a bank) to make payment to a seller. The letter of credit comes into use when there is a substantial time lag between the dispatch of goods by a seller and their receipt by the buyer. The seller, having sent the goods off, has fulfilled his part of the contract and seeks payment. The buyer, not having received the goods and being unable to inspect them, will be reluctant to pay. To overcome this difficulty the buyer and seller make arrangements to have intermediaries operating in each of the two countries involved make settlement. The buyer instructs his bank to issue a letter of credit authorizing payment to be made to the seller when the latter's part of the contract has been fulfilled (usually when the seller has dispatched the correct quantity of conforming goods). The buyer's (or issuing) bank ascertains whether or not this has been done by obtaining the cooperation of a bank in the seller's country. This bank (the "corresponding" bank), having inspected all of the relevant documents of title and bills of lading to ensure that the seller has performed, makes payment to the seller, often by means of a bill of exchange or other credit device. The document of title, bills of lading, and so forth are then mailed to the buyer. The buyer then reimburses his bank, which in turn reimburses the corresponding bank for making payment to the seller.

In no other branch of international trade have the efforts at unification of law been more successful than in that of letters of credit. In 1933 the International Chamber of Commerce in Paris published the *Uniform Customs and Practice for Documentary Credits,* which was revised in 1951, 1962, and 1983. It has been adopted by banks and banking associations in almost all countries of the world.

Loan of money. Second only to sales, the lending of money is one of the most frequent types of commercial transaction. No developed economy could exist without the credit financing of industrial investments, of commercial transactions, or of private acquisitions. A lender gives money to the borrower, who is obliged to repay it and to pay interest as well. Interest is thus the price for the utilization of the lender's money. The payment of such a price has not, however, always been regarded as permissible. For centuries the medieval Christian Church stigmatized interest as income without true work and, therefore, sinful. Religious restrictions on interest are, to this day, of great importance in the Islāmic countries. In the Socialist countries of eastern Europe and East Asia, the lending of money against interest, except through state banks, is strongly discouraged.

A loan is a contract between lender and borrower. It may consist of the immediate giving of money against the borrower's promise of repayment, or the contract may contain a promise of the lender to give the money at a future date. In the latter case it may sometimes happen that a borrower must sue an unwilling lender for performance of the promise to make the loan. As in the comparable situation of nondelivery in a sales contract, Anglo-American law (and also that of some other countries) refuses an action for specific performance and provides merely for damages, whereas most of the legal systems of continental Europe admit such an action. Specific performance of an agreement to take a loan may similarly be enforced against a borrower in most civil-law countries but not under Anglo-American law.

Interest on loans is today generally admitted. Among merchants it often must be paid even if not expressly agreed by the parties, since no merchant is regarded as willing to lend money without receiving interest thereon. Many countries fix the rate of interest to be applied in such cases. This legal rate has frequently been somewhere between 4 and 7 percent. Modern legislation sometimes establishes flexible rates, such as 2 percent above the (fluctuating) official discount rate of the country. Many countries also limit the maximum amount of interest that may be charged even if both parties have agreed on a higher rate. The maximum figure has frequently been between 6 and 12 percent, but in certain countries goes up to 30 percent or more. These "usury statutes" are likely to be circumvented by lenders who may demand considerable sums as commissions or "expenses." More flexible, but also less certain, are general laws declaring certain "usurious" transactions null and void.

Legal and maximum rates of interest

If a borrower does not repay a loan by the agreed date, he must reimburse the lender for his loss. Without even having to prove loss, the lender is at least entitled to default interest (that is, interest accruing after the due date of repayment). Some countries permit the lender to claim additional damages, whereas others exclude them.

A few countries, notably the United States, have established special rules regulating loans to consumers. This has usually been in response to abuses to which consumers have been exposed in connection with installment sales (hire-purchase agreements).

Security on loans. In the event of a borrower's bankruptcy, the lender may have to share the borrower's assets with competing creditors and may receive only partial satisfaction or even none at all. Lenders, therefore, urge borrowers to give security for the loan unless the credit standing of a specific debtor is free from any doubt. A security interest on goods (called collateral) entitles the creditor to satisfy his outstanding claim from the charged good to the exclusion of the other creditors of the borrower. Hence a security interest gives the secured creditor a right of preferential satisfaction from the goods charged with the security interest.

The demand for security on loans varies from country to country. In general, the demand is greater the more developed the credit system is. But even among countries having a comparable credit structure there are variations. Thus, certain countries, especially France, put legal obstacles in the way of modern forms of security, whereas others have recourse to various forms of personal security.

The oldest security device that is common everywhere is the pledge (or pawn). The borrower delivers the goods to be charged to the lender, who keeps them until repayment of the secured loan. This security device has become rather outmoded today and is utilized only in relatively few situations. But pawnbrokers continue to operate on a minor scale, and banks keep documents of title (such as property deeds) as security.

The decisive drawback of the pledge is the necessity of transferring the goods to be charged to the lender. Hence the borrower cannot use them for production, sale, or lease. There has thus been a trend away from the pledge to other forms of security by which the goods charged remain in the hands of the borrower. Many new devices have been introduced in the last hundred years, and they vary considerably in their operation. For want of a common descriptive name they will be referred to as "no-pledge devices." They all attempt to overcome the problem posed by the fact that third persons, relying on the outer appearance of a well-funded borrower, have no means of knowing whether or not the borrower's assets are in reality already charged in favour of another lender.

Creation of security interests

The most common method of warning third persons against existing security interests has been by their registration. Goods so charged are entered in a public register together with details about the goods themselves and the security agreement. A simpler method of giving publicity to a security interest is by marking the charged goods. This is still sometimes used in the case of cattle. Some countries also employ the method of "privileging" specific lenders. They endow the loans of certain lenders (usually publicly held or controlled banks) with a security interest or a right of preferential satisfaction. All the borrower's goods, or at least those that have been acquired by means of the loan, are automatically charged.

In the absence of any of the above three methods, various indirect techniques are usually employed. The need for sellers to retain a security interest in the goods sold until the purchase price has been paid has been particularly acute. In some, especially Latin, countries, the rules on sales provide the seller with a statutory right of preferred satisfaction. But in most jurisdictions, the seller must make his own arrangements. Since the transfer of ownership in the goods is subject to the agreement of the parties, the seller may retain his ownership in them until he has received the full purchase price. Such a "conditional sale" is recognized in many countries even without registration, since it is regarded as a modified sales transaction. If the seller himself is using credit to finance his credit sales, the financer can usually be secured by transferring to him the seller's retained ownership. In some countries, including Great Britain, the so-called hire-purchase method is widespread, especially in sales to consumers. The seller retains ownership but surrenders the goods that the buyer intends to acquire on hire to him against a down payment and a monthly rental. If in due course the rental payments accumulate to the sale price, ownership is transferred to the purchaser. Here again registration is usually not required since the transaction is cast into the form of a lease.

In the case of lenders who are not at the same time sellers it has been more difficult to find adequate security devices. One of the most successful methods has been developed in close analogy to hire-purchase sales transactions. The borrower transfers ownership in the goods financed by the loan to the lender but retains them in his possession by means of a lease agreement between him and the lender. After repayment of the secured loan the borrower reacquires title from the lender.

The disposition of charged goods

Under modern economic conditions it is rarely feasible to deprive the borrower of charged goods, as the rise of no-pledge security interests demonstrates. But the goods frequently do not even rest in the borrower's hands. This is especially likely to be the case if the borrower is a trader; he will probably want to sell the goods that he has charged. A manufacturer may similarly wish to replace charged machinery. In these instances the need arises to allow the borrower the desired disposition of the goods and at the same time to maintain the lender's security interest. A number of legal systems, however, do not yet recognize the legitimate interests of both parties in this situation. They prohibit any disposition by the borrower and may not admit a security interest in goods that remain in the borrower's hands for purposes of resale. This problem has been solved in the United States and Great Britain, although on slightly different lines in the two countries. In the United States the original registration may provide that the security interest is to extend to the proceeds from the disposition of the goods or to products of the charged goods. In Great Britain the right to a "floating charge," granted against its assets by a borrowing company to a lender, has the same effect.

A security interest proves its legal value when under attack by third parties. If it is to fulfill its function of guaranteeing to the lender preferential satisfaction of his claim, the charged goods must be immunized as much as possible against the rights of other persons. The third party is frequently a person who has unknowingly purchased charged goods: borrowers in financial straits may be unable to resist the temptation of selling charged goods left in their hands to a third person without the consent of the lender and without making the proceeds of the sale available to him. Most countries tend to protect the buyer, provided he neither knew nor ought to have known of the existing security interest. If the charged goods are marked, the buyer can hardly claim to have purchased in good faith. But mere registration does not usually give the buyer sufficient notice, since sales transactions cannot be burdened by requiring the buyer to search a register in a distant place.

The borrower's other creditors are also likely to have an interest in charged goods. But creditors, as distinct from buyers, are usually expected to search existing registers of charged goods. If they have neglected to do this, they must suffer the consequence of being subordinated to the lender's security interest.

Enforcement of security interests

Failure of repayment. If the borrower does not make payment after the secured loan has fallen due, the lender may pursue two different courses. He may enforce his claim for repayment before the courts just as any other creditor or he may enforce his preferred position as a secured lender. The rules to be followed in enforcing a security interest differ considerably from country to country and even within a country according to the type of security interest involved. Very often the lender must sell the charged goods by public sale; occasionally he is permitted to acquire the charged goods himself. If the proceeds of a sale exceed the amount of the secured loan, the surplus must be paid to the borrower, whereas the borrower remains liable for any deficit. All legal systems frown upon clauses that permit a lender to acquire the charged goods automatically on the borrower's failure to pay.

The rules on security interests are still strongly national in character. The need for unification has, except in a few specialized areas, not been very urgent. This is largely because, in the great bulk of international sales transactions, the seller, wherever necessary, may secure himself by insisting on payment by letter of credit. But of some international concern was the question of the protection of security interests in those means of transportation that move constantly from one country to another. Two international conventions on security interests in ships and aircraft have, therefore, been concluded. They do not provide uniform rules on security interests but merely guarantee that an interest validly created in one contracting state will be recognized in any other contracting state. The number of countries that have adopted these conventions is, however, limited. (U.M.D.)

Bankruptcy and insolvency

Bankruptcy laws were enacted to provide and govern an orderly and equitable liquidation of the estates of insolvent debtors. This purpose has remained an important aim of bankruptcy legislation since the Middle Ages. Because in the past bankruptcy was coupled with the loss of civil rights and imposition of penalties upon fraudulent debtors, the designation bankrupt came to be associated with dishonesty, casting a stigma on persons who were declared bankrupts. Eventually, however, bankruptcy legislation was extended to provide procedures for the adjustment of debts so as to avoid liquidation and for the rehabilitation of insolvent debtors. Modern bankruptcy laws, therefore, include detailed provisions for preventive compositions, arrangements, or corporate reorganizations of various types. In fact, the salvage of an enterprise in financial difficulties has become the principal focus of insolvency legislation with particular concern for the maintenance of employment opportunities and the protection of members of the labour force.

In addition, the bankruptcy laws of England, the United States, and the British Commonwealth nations traditionally came to include provisions for the unpaid portions of debts incurred prior to bankruptcy in order to give honest but unfortunate debtors a new start in life. The bankruptcy laws of the European and Latin-American countries, by contrast, did not have such provisions. In the late 20th century, however, legislation in some of these countries (*e.g.,* Argentina and France) provided for the discharge of the unpaid portion of pre-bankruptcy creditors under certain conditions.

Since bankruptcy laws aim at the liquidation or rehabilitation of insolvent estates, bankruptcy proceedings involve all nonexempt assets of the debtor, and all creditors entitled to share in the proceeds of the liquidation or in the adjustment of their claims are called to participate. Accordingly, bankruptcy proceedings are viewed as general or universal collection procedures as distinguished from individual collection remedies available to particular creditors for the enforcement of their claims.

HISTORY OF BANKRUPTCY LAW

Early developments

Modern bankruptcy law has been formed from a number of distinct historical strands. In ancient Roman law an unpaid judgment creditor could have the debtor's estate sequestered (*missio in bona*) and sold for the benefit of all creditors (*venditio bonorum*). Proceedings of this type caused loss of civil rights. To alleviate this hardship a debtor was given the privilege of relinquishing voluntarily his assets to his creditors by petitioning a magistrate (*cessio bonorum*).

During the Middle Ages both institutions underwent a revival and development. The medieval Italian cities enacted statutes dealing with the collection and distribution of the assets of debtors, especially merchants, who had absconded or fraudulently caused insolvency. Such bankrupts (*rumpentes et falliti*) were subjected to severe penalties, and their estates were liquidated. In addition, medieval Spanish law restored the judicial *cessio bonorum.* The Siete Partidas, a codification published by authority of Don Alfonso X the Wise, king of Castile and León, during the second half of the 13th century, contained detailed provisions relating to insolvent debtors, applicable to merchants and nonmerchants alike, enabling them to secure a voluntary liquidation of their assets under judicial supervision. An unpaid creditor could insist on either payment or assignment of his estate by the debtor to all creditors.

Laws dealing with the property of absconding and fraudulent debtors, modeled after the statutes of the medieval Italian cities, spread throughout western Europe. Provisions of this type were adopted in the commercial centres of France, Brabant, and Flanders during the 15th and 16th centuries. The customs of Antwerp, printed in 1582, contained comprehensive rules on the treatment of bankrupts and their estates. The emperor Charles V, as count of Flanders, inserted stringent provisions for the repression of bankruptcies in his Decree for the Administration of Justice and Good Order of 1531. There can be no doubt that the first English "acte againste suche persones as doo make Bankrupte," passed in 1542/43, was inspired by the northern European models, as the title reproduces the Flemish expression. It governed proceedings instituted against absconding or concealed debtors. It was replaced by a more detailed act of 1571 that applied only to merchants and other traders. Voluntary proceedings were not provided in England until 1844 and not in the United States until 1841.

In France, national rules on insolvency and bankruptcy were inserted into the Ordonnance du Commerce of 1673. It regulated both voluntary assignments for the benefit of creditors made by merchants (Title X) and the proceedings and effects flowing from bankruptcy (Title XI). It was interpreted to restrict bankruptcy proceedings to merchants only, and the laws of many other countries followed the French lead. Thus in Spain the limitation of bankruptcy to merchants was adopted by the Ordinances of Bilbao, which were sanctioned in 1737 and subsequently applied in Latin America, especially Argentina.

The restriction of bankruptcy legislation to persons engaged in commerce created a need for liquidation proceedings applicable to other debtors. As mentioned above, the Siete Partidas contained provisions for voluntary liquidation proceedings applicable to all classes of debtors. On that basis a Spanish jurist of the 17th century, Salgado de Somoza, elaborated detailed rules for the initiation and conduct of voluntary liquidation proceedings, which were styled "concourse of creditors." His tract, entitled *Labyrinthus Creditorum,* influenced the course of Spanish law and also had great impact on the common law of the German states. As a result, Spanish law developed two classes of liquidation proceedings, one for merchants and one for nonmerchants. Spanish law in that respect was the model for the legislation in Portugal, Argentina, Brazil, and other Latin-American countries. Other nations, including Austria, Germany, England, the United States, and nations influenced by English laws, brought both merchants and nonmerchants under their bankruptcy laws. More recent laws in Latin America (*e.g.,* in Argentina and Peru) likewise have established a unified system. France, Italy, and a few countries in Latin America, however, do not provide true insolvency proceedings for ordinary debtors.

Avoidance of bankruptcy

The dire consequences of bankruptcy for the debtor, such as the loss and liquidation of his assets, criminal penalties, and loss of civil rights, resulted in the need for procedures avoiding such sanctions. A remedy was found in the right of a deserving debtor to reach an agreement for an extension or reduction of his debts with a majority of his creditors that was binding on dissenters. The cradle of this institution was again the statutes of the medieval cities. Provisions to that effect were also contained in the Siete Partidas. In England, similar procedures were developed by the Privy Council through bills of conformity, but this practice ended with the abolition of the council's civil jurisdiction in 1641. In France the Ordonnance du Commerce of 1673 recognized majority compositions as a legitimate means of handling the estates of insolvents without liquidation. The Commercial Code of 1807, however, and following it the laws of other countries, restricted them to a method of terminating rather than preventing bankruptcy proceedings. Preventive compositions were reintroduced as legitimate means of dealing with embarrassed or insolvent estates only during the second part of the 19th century; they are now recognized in most countries as important devices for economic rehabilitation.

At one time all bankrupts were considered defrauders and criminals. They were subjected to severe social and professional sanctions, including even a degrading form of dress. In recent times, however, great efforts have been made to remove the disgrace attached to bankruptcy. Even the terms bankrupt and bankruptcy (or their equivalents in other languages) are used less and less frequently in the statutory language. Modern French legislation, for example, totally suppresses the traditional term *faillite* as the name of liquidation proceedings and restricts it to special procedures entailing the imposition of disqualifications on insolvents guilty of commercial misconduct.

LIQUIDATION OF THE ESTATES OF INSOLVENT DEBTORS

Most nations with private-enterprise economies have legislation providing for liquidation of hopelessly insolvent estates. Socialist countries treat financial difficulties of their economic organizations in a manner aiming at rehabilitation rather than elimination. By the late 20th century, however, many countries not considered Socialist likewise were resorting to liquidation of enterprises only as an ultimate measure and instead favoured techniques to rescue distressed enterprises, sometimes by direct financial intervention. Conversely, Yugoslavia and, more recently, Hungary have enacted legislative measures in the nature of bankruptcy laws. The People's Republic of China has introduced draft legislation governing the liquidation of insolvent enterprises by way of experiment.

Liquidation proceedings are often referred to as "straight" bankruptcy, in contradistinction to other insolvency proceedings aiming at arrangements or reorganizations. Recent insolvency laws, such as those enacted in Argentina and France, provide for a unified procedure in which liquidation is decreed only after the possibility of reorganization has been found not to exist or an attempted reorganization has failed.

Differences in coverage of laws

Persons subject to judicial liquidation of their estates. Bankruptcy or insolvency laws vary considerably in their applicability to particular classes of persons. The West German act and, following its example, the Austrian and Japanese acts extend bankruptcy proceedings to all natural and legal persons, whether or not they are engaged in commerce and without differentiating between petitions by the bankrupt himself or by creditors. In the United States, individuals, whether merchants or nonmerchants, as well as private corporations, with the exception of certain financial institutions, are subject to the Bankruptcy Code. Involuntary petitions, however, cannot be filed against farmers and nonprofit corporations. Moreover, proceedings for debt adjustment of individuals cannot be initiated by creditors. Canada likewise applies its act to individuals and corporations. It excludes, however, certain financial institutions and nonbusiness corporations in general. In England the former Bankruptcy Act covered only individuals, whether merchants or not. Registered companies were liquidated under the winding-up provisions of the Company Law. A great number of the provisions of the Bankruptcy Act, however, were made applicable in such proceedings. The dual system still governs in Australia, New Zealand, and India. A number of nations, following the model of the French law of 1838, extend their bankruptcy laws only to persons qualifying as merchants or engaging in trade but do not differentiate between individuals and corporations. To that class belong the bankruptcy laws of Italy (with the exception of small enterprises), Spain, Portugal, Switzerland, and a number of Latin-American countries, including Bolivia, Brazil, Colombia, Mexico, and Venezuela. Argentina, Chile, and

Peru, however, follow the German pattern and subject to their bankruptcy laws all individuals and corporations, whether merchants or not. A number of the countries that restrict bankruptcy to merchants have, however, inserted provisions for insolvency proceedings governing nonmerchants in their codes of civil procedure. France extends its new insolvency law (1985) to merchants, artisans, and all legal persons even if not merchants.

Persons entitled to initiate liquidation proceedings. Modern bankruptcy laws provide for the initiation of liquidation proceedings upon petition by either the bankrupt himself or his creditors. There are differences as to the number of creditors who must join in a creditor's petition. A great number of laws are satisfied with a petition by a single creditor regardless of the amount of his claim, the total number of creditors, or the amount of the outstanding indebtedness, so long as the debtor is unable to meet his current payments or has committed a so-called act of bankruptcy. Petition by a single creditor suffices according to the law of Germany (1877) and, following it, of Japan; the laws of Austria, France, Italy, Portugal, Spain, and Switzerland; and the laws of such Latin-American countries as Argentina, Brazil, Chile, and Mexico. In Austria at least one other creditor must exist, although he need not join in the petition; but the sufficiency of one unpaid creditor is provided expressly in the Chilean and Mexican acts.

A somewhat different regime exists in the common-law countries. In England, Canada, Australia, and New Zealand, as well as India, a single creditor may be a petitioner if the unsecured part of his claim equals or exceeds a specified amount. Otherwise, other creditors must join until the aggregate amount of their claims equals the requisite sum. In the United States in 1978, the unsecured indebtedness of the bankrupt had to be at least $5,000, and a petition by three or more creditors with unsecured claims aggregating at least that amount was required if the total number of creditors was 12 or more; otherwise one or two creditors with such claims totaling at least that amount might be petitioners.

In some countries the initiation of liquidation may also be decreed by the court ex officio or upon petition by public officials. Ex officio action by the court is provided, for example, in Italy, France, and Mexico. Moreover, these countries and, in addition, East Germany and Portugal authorize public officials to file petitions for the liquidation of bankrupt estates. In a number of countries, following in that respect the traditional French approach, a bankrupt debtor is under a duty to file a petition, and his initiation is not left to his own judgment as in the common-law countries.

Substantive prerequisites of liquidation proceedings. The bankruptcy laws of the various nations differ materially as to the definition of the substantive grounds for the institution of insolvency proceedings, especially those initiated by creditors. The majority of laws employ general formulas, such as cessation of payments (Argentina); impossibility to meet current indebtedness with disposable assets (France); lack of ability to pay all debts, taking account of contingent and prospective liabilities (England); general nonpayment of debts, not subject to a bona fide dispute, as they mature (United States); inability to pay or excess of liabilities over assets (West Germany); or inability to satisfy regularly one's liabilities (Italy). Some of the common-law countries, including Australia, Canada, India, and New Zealand, require a creditor's petition upon commission by the debtor, within a specified period prior to the petition, of one or more "acts of bankruptcy" or "acts of insolvency" listed in the respective statutes. These acts, which vary among statutes, include public manifestations of insolvency as well as conduct that endangers the collectibility of debts or entails preferential treatment of certain creditors by an insolvent. Some jurisdictions have mixed systems. Liquidation is decreed if the debtor has either resorted to cessation of payments or committed specified acts manifesting insolvency. Laws of that type apply in Spain, Portugal, Brazil, Chile, and Mexico. In some of these laws (*e.g.,* that of Mexico) the commission of these specific acts raises merely a presumption of a cessation of payments. In Brazil and Chile even a single default in the payment of a liquid and exigible indebtedness warrants proceedings if the obligation has remained unpaid after demand. In Switzerland institution of liquidation proceedings likewise can be based on cessation of payments or on the commission of other specified acts of bankruptcy.

Acts of
bankruptcy

Assets subject to liquidation proceedings. One of the most important aspects of bankruptcy legislation is the determination of the assets to be seized and sold for the purpose of distributing the proceeds among the creditors. Various legal systems have vastly different approaches. The disparities relate mainly to the status of assets acquired by the bankrupt subsequent to his adjudication or conveyed away by him prior to that date.

The
bankrupt's
estate

In West Germany all nonexempt assets belonging to the bankrupt at the date of the adjudication form the bankrupt estate. Assets that are no longer owned by the bankrupt at the time of the adjudication are not included in the bankrupt estate unless their sale, transfer, or other disposal is voidable under special rules permitting the avoidance of fraudulent or preferential transactions. The Bankruptcy Code of the United States follows a similar approach, except that the "date of cleavage" is not the date of the adjudication but the date of the filing of the petition. Postpetition acquisitions are part of the bankrupt estate under the U.S. law only if they constitute narrowly defined "windfalls," such as inheritances or bequests settled on the bankrupt within six months from the filing date. On the other hand, many other bankruptcy laws include within the estate subject to distribution all nonexempt assets acquired after the adjudication and during the period of the proceedings. Thus the English Insolvency Act, 1986, provides that in individual bankruptcies the bankrupt's estate comprises all nonexempt property owned by the debtor on the day on which the bankruptcy order is made and any property claimed by the trustee (the person charged with the administration and liquidation of the bankrupt's estate) that has been acquired by or devolved upon the bankrupt since that time and until the date of his discharge. Similar provisions exist in Canada, Australia, and New Zealand, except that in Australia and New Zealand the date for determining what is property of the bankrupt estate relates back to (*i.e.,* is retroactive to) the date of the earliest commission within a specified period of an act of bankruptcy by the bankrupt. This "relation-back" theory, however, is qualified by numerous exceptions.

In the majority of civil-law countries that follow the traditional French model the bankrupt estate likewise includes all nonexempt property owned by the debtor at the date of the adjudication and all property acquired during the course of the proceedings until the close of the case or the rehabilitation of the debtor. Subject to variations in details, laws of that type operate in, for example, Argentina, Austria, Brazil, Italy, Portugal, Spain, and Switzerland. Some of these laws (*e.g.,* those of Chile and Italy) make special exceptions for future earnings to the extent that they are necessary for the debtor's maintenance. The 1985 French law on economic rehabilitation and liquidation also includes in the bankrupt estate all property acquired on whatever grounds until the close of the proceedings. With the exception of Austria, Italy, and Switzerland, the countries listed above retract the effective date of adjudication to the date of the cessation of payments.

The insolvency or bankruptcy laws of England and other countries that follow the English model give title to the property forming the bankrupt estate to the trustee or assignee in bankruptcy. In the United States the estate is a separate legal entity represented by a trustee. In the other countries liquidation proceedings divest the bankrupt of his power of administration and disposition, but he retains the title to the assets of the estate. The estate, however, forms a separate unit, often called the mass.

The
bankrupt's
estate

As indicated above, a number of bankruptcy laws contain a relation-back effect, dating the bankrupt estate as of the time of the petition, the earliest commission of an act of bankruptcy, or the cessation of payments. The model for that approach in the civil-law countries whose laws were based on the French codes was the French Commercial

Code of 1807. It required the decree of adjudication to fix the date of the cessation of payments and provided that transactions by the bankrupt after that time and before the adjudication (called the critical, or suspect, period) did not affect the estate. In 1838 France replaced the general relation-back theory with a catalog of transactions, mostly various types of gratuitous or preferential transfers, that were rendered ineffective against the mass if made during the critical period. The law of 1985 in essence retains that approach. Most countries limit the duration of the period of possible retroactivity. Spain is one of the countries still adhering to the general invalidation of transactions following the cessation of payments. Italy abolished the relation-back effect in 1942.

Preferences. One of the cardinal principles governing the liquidation of insolvent estates is the equal treatment of creditors—the classical *par condicio creditorum.* Debtors on the eve of bankruptcy, either on their own volition or under pressure, may accord preferential treatment—by way of payment or security—to certain creditors. The bankruptcy laws of most, if not all, countries therefore contain rules aiming at the reintegration of the bankrupt estate through avoidance of preferences given after insolvency or cessation of payments or even earlier. These provisions are included in, in lieu of, or cumulative with a general relation back of the effects of the adjudication. Again, the laws of the different countries vary greatly with respect to the elements, especially those of a subjective character such as intent on the part of the debtor or knowledge of the debtor's financial status on the part of the creditor, that must be present to make a transfer voidable as a preference.

In the United States, except in the case of insiders, the preference period is 90 days before the date of the petition; transfers to existing creditors made within that period that have a preferential effect are voidable, unless the debtor was neither insolvent at the time of the preferential transfer nor rendered insolvent thereby, or the payment was made in the regular course of business. In England a preference given by an insolvent, or an individual rendered insolvent thereby, to a creditor other than an associate of the debtor is voidable if it was made within six months prior to the filing of the bankruptcy petition and was motivated by the desire to give a preference. In Canada, in cases other than preferences to creditors related to the debtor, the preference period is three months prior to the receiving order (*i.e.,* the order initiating the proceedings). The transfer must have been made or suffered by an insolvent with a view to giving the creditor a preference; but this element is presumed to exist if the transfer in fact had a preferential effect. In Australia the preference period commences six months prior to the petition. A transfer within such preference period by the bankrupt to a creditor that gives the creditor a preference is voidable if the debtor was insolvent at that time, unless the creditor had no reason to suspect that the debtor was insolvent at that time and that the transfer would give him a preference. The New Zealand law differentiates between preferences made by an insolvent with the view of giving a preference and preferences made without such purpose. In the first alternative the preference period is two years prior to the date of adjudication, in the second it is one month.

Presumption of intent

The French law of 1985, preserving vestiges of the traditional relation-back theory, invalidates payments of debts made after their due date and the date of the cessation of payments if the recipient had knowledge of the cessation of payments. In addition, anticipatory payment of unmatured debts, payment of debts made by other than ordinary means of payment, or the grant of security interests for preexistent debts are invalid if made after the cessation of payments, regardless of the creditor's knowledge. Rules of the latter type also govern in Latin-American countries (*e.g.,* Argentina, Brazil, Chile, and Mexico). In Argentina the payment of matured debts with ordinary means may be voidable if the creditors knew of the cessation of payments. The laws of these countries vary as to the outer limits of the suspect period, ranging from two years in Argentina and Chile to 60 days in Brazil.

In the civil-law countries that do not adhere to a general relation-back doctrine, such as West Germany, Italy, or Portugal, preferential transfers to creditors prior to an adjudication may be voidable, but the governing laws vary materially as to the relevant time frames and types of preferences. In West Germany transfers by the debtor granting a security interest or satisfaction to which the creditor at that time was not entitled are voidable if made within 10 days prior to or after a cessation of payments or filing of a bankruptcy petition, unless the creditor shows that at that time he knew neither of the cessation of payments or petition nor of a preferential intent on the part of the debtor. The voidable act may not precede the adjudication by more than six months. In addition, security interests or satisfaction obtained by the creditor after a cessation of payments or filing of a petition are voidable if the trustee shows that the creditor knew of the cessation of payments or petition. In that case the date of the adjudication is irrelevant. Similar rules governing preferences apply in Austria. Voidability is provided if the preference was obtained within 60 days prior to or after the petition or within one year prior to the adjudication after insolvency. Moreover, in the case of transfers to the creditor in conformity with an obligation, the trustee must prove the preferential intent of the debtor and knowledge thereof by the creditor, whereas in the case of payments made or security interests granted not in conformity with an existing obligation, the only defense is lack of preferential effect. In Italy anticipatory payments of debts that fall due only on the day of or after the adjudication are *ipso jure* (by the law itself) invalid if made within two years prior to the adjudication. Payments, effected by normal means, of debts that were payable at the time of such payment are voidable as preferences provided that the payment was made within one year prior to the adjudication, while the debtor was insolvent, and the trustee shows that the creditor knew the debtor's state of insolvency. Payment of matured debts made by other than normal means and security interests granted for preexisting unmatured debts are voidable if made within two years prior to the adjudication unless the creditor shows that he was ignorant of the debtor's state of insolvency. In Switzerland voidable preferences consist in the grant, without prior obligation to that effect, of security interests for existing debts; satisfaction of debts by other than usual means of payment; or premature payment of debts, if the respective acts were committed within six months prior to the adjudication, while the debtor was insolvent. The creditor can defeat the avoidance by showing lack of knowledge of the debtor's financial condition. In addition, all acts committed with the intent to prefer a creditor are voidable for five years after their execution, if the intent was recognizable by the creditor.

Creditors entitled to share in the distribution of the estate. One of the principal objectives of bankruptcy is the distribution of the proceeds of the estate among the creditors. The designation of the categories of claims entitled to share in such distribution, the order of distribution, and the regulation of the procedures for their establishment form an important ingredient of modern bankruptcy legislation. At the outset a distinction must be made between secured and unsecured creditors. The recognition and assertion of security interests is subject to rules that aim at the separate satisfaction of the secured party from the proceeds of the collateral. To the extent that a secured creditor holds a claim that exceeds the value of the collateral—a so-called undersecured creditor—the creditor must comply with the rules governing unsecured debts. Many laws also require timely claim of security interests to permit orderly distribution of the estate. Moreover, the separate enforcement of security interests may be barred by the institution of bankruptcy proceedings. Security interests may be voidable as fraudulent or preferential transfers or because of inconsistency with overriding policies of the bankruptcy legislation. Except for these limitations, a security interest will remain unimpaired by bankruptcy and provide priority over creditors who do not possess a right of satisfaction from the specific property. The French law of 1985, however, subordinates all types of security interests to the rights of creditors originating during the period between the order instituting the procedure for economic rehabilitation or liquidation and the order of liquidation, and there is a general complaint against the consumption of the assets by secured creditors.

Provable claims

Generally speaking, "bankruptcy debts" or "provable claims" are based on transactions or occurrences prior to the date of bankruptcy. Creditors whose claims result from acts in, or expenditures for, the administration of the bankrupt estate are not creditors of the bankrupt and are paid before the latter. Post-bankruptcy debts incurred by the debtor are not payable from the estate. The line that divides provable post-bankruptcy claims from non-provable post-bankruptcy claims varies. In England, Australia, and Canada debts provable in bankruptcy (bankruptcy debts) include any debt or liability to which the bankrupt is subject at the date of the bankruptcy or to which he may become subject before his discharge by reason of an obligation incurred before the date of bankruptcy. In England liabilities accruing after the debtor's discharge on an obligation incurred before bankruptcy are also provable. The date of bankruptcy within the meaning of these laws is that of the adjudication or making of the sequestration or receiving order, respectively. In the United States the date of the petition controls the determination of provable debts, but "gap creditors" are included if the adjudication follows an involuntary petition. In continental European and Latin-American countries and countries influenced by their laws, such as Austria, France, West Germany, Italy, Portugal, Spain, Argentina, Brazil, Chile, and Japan, creditors entitled to share in the estate are likewise those whose claim originated prior to the adjudication, while subsequent creditors are relegated to assets not entering into the estate.

Under most modern bankruptcy laws provable debts include all types of pre-bankruptcy obligations, whether matured or unmatured, liquidated or unliquidated, unconditional or contingent. A provable unliquidated liability may be based on contract or tort, including personal injury as well as damages to property. This also holds true for most common-law countries, in particular England, the United States, Canada, and New Zealand. In Australia, however, demands for unliquidated damages arising other than by reason of contract or breach of trust are not provable. In Brazil creditors with unliquidated claims are entitled to the setting aside of a reserve. If at the time of proof the amount of the liability is still undetermined, many laws provide that it may be estimated.

In principle, bankruptcy laws aim at equality among creditors. Nevertheless, for social or fiscal reasons bankruptcy laws recognize or grant preferential rights to certain categories of claims. These rights are not security interests in particular assets but priorities in the distribution of the proceeds from their sale. These priorities may attach either to the proceeds from the entire estate or only to those from certain classes of assets. Some bankruptcy laws establish complex hierarchies of priority. The most common general priorities relate to tax claims and claims of or for the benefit of the labour force. In some countries labour claims not only have priority over all other provable claims but outrank even the rights of secured creditors. Examples of the latter approach are the laws of Brazil and France.

Discharge of debts. An English statute of 1705 provided for a discharge (release) of all debts owed by the bankrupt and due at the time of bankruptcy provided the bankrupt had faithfully complied with his statutory duties. Since that time relief of the honest but unfortunate debtor from his provable debts has become one of the main objectives of bankruptcy legislation in countries influenced by the English system. The right to a discharge is not unqualified and may be forfeited by the commission of acts that the particular legislation considers as meriting this sanction. Moreover, certain debts are excepted from the operation of the discharge, as, for example, liabilities for support under the governing family-law provisions or for certain types of personal injury.

In England discharges are automatic after the expiration of specified periods, usually three years, unless the court has suspended or conditioned the running of the period. In the United States a discharge requires a court order, but the debtor is entitled to a discharge unless he has engaged in conduct defined by law that disqualified him from receiving it and the trustee or a creditor has objected to the discharge. There is no statutory waiting period.

In Australia and New Zealand a debtor is automatically discharged three years after the date of adjudication, unless an objection to the discharge has been entered by the trustee or a creditor. In addition, the bankrupt may obtain an earlier discharge by court order upon application, which may be subject to conditions. Conditional discharges also may be ordered under Canadian law. Provisions for discharge exist in South Africa and in an increasing number of civil-law countries. Japan introduced discharge provisions modeled after U.S. law in 1952. Brazil, Argentina, and, most recently, France adopted provisions for the extinction of debts to the extent that they remain unsatisfied by dividend payments. In France that rule is absolute; in Argentina and Brazil it is subject to specified conditions and qualifications.

Roles of the court, the administrator, and the creditors in bankruptcy proceedings. Since insolvency proceedings aim at a judicially approved or supervised rehabilitation or liquidation of the estate of an insolvent, the proceedings must be initiated in a court having the necessary judicial powers. The bankruptcy laws of different countries, however, vary considerably with respect to the jurisdiction of the bankruptcy court over the phases of a case and the relative roles assigned to judicial officers, administrators, and creditors in the conduct of a case.

Jurisdiction of the bankruptcy court

During the formative period of bankruptcy law in the 18th and 19th centuries the courts developed the doctrine of the "force of attraction" of bankruptcy proceedings, resulting in the concentration in the bankruptcy court of all litigation relating to the creditors or the assets of the estate. Modern statutes still adhere to that idea, but with considerable variation. The laws of England, the United States, Australia, Canada, and New Zealand contain very broad jurisdictional grants to the bankruptcy courts relating to the collection and administration of the property of the estate and the rights of secured and unsecured creditors. A number of civil-law countries, following French and Spanish models, provide equally extensive powers to the bankruptcy courts. In some of these countries, however, the jurisdiction of special labour courts limits the judicial powers of the bankruptcy courts. In West Germany the settlement of individual controversies, such as the determination of contested claims or the rights of third parties to or in property claimed as part of the bankrupt estates, is left to the regular courts. In Austria, however, the bankruptcy court has exclusive or concurrent jurisdiction in disputes of that type.

In most countries many functions of the bankruptcy court are conferred upon special judicial officers who may be either actual members of the judiciary (as in France) or judicial officers without full judicial status. In England registrars of the courts vested with insolvency jurisdiction, especially the bankruptcy registrars of the High Court, have jurisdiction over the initiation of insolvency proceedings and a long catalog of matters not requiring hearing in open court. Similar rules apply in Canada and New Zealand. In the United States special bankruptcy judges may exercise most judicial functions involved in the conduct of a case. Since these judges are not appointed for life and with Senate approval, some functions are reserved to the regular judiciary. Similar reasons limit the delegation of judicial functions in Australia. In West Germany, conversely, a wide range of judicial functions in bankruptcy are delegated to court registrars (called *Rechtspfleger*) who perform judicial functions outside the regular judiciary.

Growing power of the courts

Creditors have traditionally played an important part in the conduct of insolvency cases. In most countries, however, they are no longer the dominant parties. The courts, assisted by official administrators, have emerged as the key figures. Under the 1985 French law even the creditor's representative is a functionary of the administration of justice, appointed by the court from an approved panel. In England the system of creditors' control that existed from 1706 to 1831 and from 1869 to 1883 was replaced in the latter year by a system of joint control, or creditor participation, a scheme that in attenuated form is retained by the Insolvency Act of 1986. The administration of bankrupts' estates is undertaken by an official receiver—an office established in 1883—unless a general meeting of creditors or the secretary of state for trade and industry, upon reference by the official receiver, appoints another person qualified to act as an insolvency practitioner to the post of trustee. In cases of summary administration (a simplified procedure that may be ordered by the courts if unsecured debts fall below a certain level) the appointment may be made by the court. Official receivers exist also in Australia and Canada. In Australia the official receiver functions as official trustee only when there is no registered trustee. Since 1981 a registered trustee is the initial trustee if he consents to the appointment prior to the adjudication. He may be removed by the court upon petition by a creditor. The official trustee may be replaced by a registered trustee upon resolution by a creditors meeting. In addition the creditors may appoint a committee to advise and superintend the trustee. In Canada official receivers, appointed by the governor in council, are not trustees of the insolvent estate but exercise supervisory authority. The initial trustee is appointed by the court from a list of licensed trustees but is subject to being replaced by another licensed trustee by means of a special resolution adopted at any meeting of the creditors. In New Zealand the administration of insolvent estates is entrusted to official assignees, appointed under the State Services Act. The creditors may appoint a committee to assist the assignee in the exercise of his functions. In the United States the court appoints an interim trustee to serve until the creditors elect another qualified person to serve as trustee in the case; if the creditors fail to act, the interim trustee remains in office. By way of experiment, administrative officers, called U.S. trustees, are appointed for some districts to serve as interim trustees and, if the creditors make no other election, as trustees. In Chile trustees are licensed and supervised by a governmental agency. In the order of adjudication the judge appoints an interim trustee who remains in office if his appointment is ratified in the first meeting of the creditors, though they may appoint another licensed person as definitive trustee. Chile's bankruptcy act of 1982 has restored creditors' control to some extent. In Switzerland bankruptcy offices are established for each of the bankruptcy districts created by the individual cantons. The bankruptcy office is in charge of the progress of the proceedings and acts as trustee unless the creditors select one of their choice. In Italy the bankruptcy judge appoints the trustee and selects a committee from among the creditors. The committee has merely advisory and supervisory functions. A meeting of creditors is summoned only for the determination of provable debts.

REHABILITATION AND REORGANIZATION OF INSOLVENT ESTATES

The principal focus of modern insolvency legislation no longer rests on the liquidation and elimination of insolvent estates, but on the remodeling of the financial and, if necessary, organizational structure of a debtor in economic difficulties so as to permit the continuation of the economic activities. Procedures for the conclusion of binding preventive accords with creditors, requiring approval by a qualified majority of them and confirmation by the courts, were provided by legislation in a number of countries during the second half of the 19th century. Economic crises during the 20th century gave rise to further legislative measures, providing a breathing spell for the debtor by means of a postponement or reduction of his liability, coupled if necessary with a change in ownership. Thus a German law of 1916 provided for avoidance of bankruptcy and executions by a procedure placing the debtor, upon his petition, under judicially supervised management. Similar procedures were introduced subsequently into the companies acts of South Africa and Australia and are now incorporated into the English Insolvency Act, 1986, as a new avenue of relief for debtors threatened with insolvency (administration orders procedure). In addition, the laws of England, Canada, Australia, and New Zealand outline procedures for preventive composition and schemes of arrangement with creditors by companies or individual debtors. Procedures for preventive accords also exist in many civil-law countries—*e.g.,* Austria, West Germany, Italy, Portugal, Spain, Argentina, Brazil, Chile, and Mexico. Italy has enacted special legislation for the extraordinary management of large enterprises in economic difficulties. Usually judicial compositions affect only the rights of unsecured creditors. Secured creditors must specifically and individually assent to any modification of their rights. In the case of corporate debtors, however, effective restoration of the viability of the enterprise may require much more drastic measures. Therefore, the United States, Japan, and, more recently, Argentina and Chile have enacted laws establishing procedures to permit the formulation and judicial confirmation of reorganization plans, the provisions of which may include elimination of ownership rights and significant curtailments of the rights of secured creditors. Austria reformed its composition act in 1982 by adding a preliminary procedure, resembling the old German management supervision order, to facilitate voluntary reorganization, especially refinancing. The most comprehensive legislation aimed at the salvage of distressed enterprises is the French insolvency legislation of 1985. This law provides a unified procedure establishing a mandatory period of observation to determine whether the insolvent enterprise can be rescued, if necessary at the expense of the owners and existing creditors, or whether liquidation is unavoidable.

Preventive accords

INTERNATIONAL ASPECTS

Normally the existing bankruptcy laws do not differentiate between foreign and domestic creditors in proceedings involving the estates of residents, at least if reciprocity exists between the countries of the parties involved. Provisions to that effect exist in West Germany and Japan and, by implication, in Italy. A number of Latin-American countries, however, give priority to local creditors if there are concurring bankruptcies (simultaneous proceedings in more than one country) involving the same debtor. Equally controversial is the extraterritorial effect of releases resulting from compulsory compositions or discharges in bankruptcy.

Bankruptcy treaties

Because of the difficulties arising from multiple bankruptcies or the principle of territoriality of bankruptcy legislation, some countries have regulated that subject among themselves by regional conventions. Among the Latin-American countries are three treaties, binding different parties: the two treaties of Montevideo on International Terrestrial Commercial Law, concluded in 1889 and 1940, respectively, and the treaty of Havana on Private International Law (the so-called Bustamante Code) of 1928. The five Scandinavian countries concluded the Copenhagen Convention on bankruptcy on Nov. 7, 1933. In the late 1980s the members of the European Economic Community were studying the draft of a bankruptcy treaty, proposed in 1980, which superseded a preliminary draft of 1970. In addition, there exist a number of bilateral treaties between different nations on the subject—*e.g.,* the West German–Austrian bankruptcy treaty of 1979, which entered into force in 1985. (S.A.Ri.)

Labour

Labour law is the varied body of law applied to such matters as employment, remuneration, conditions of work, trade unions, and industrial relations. In its most comprehensive sense the term also includes social security and disability insurance. Unlike the laws of contract, tort, or property, the elements of labour law are less homogeneous than the rules governing a particular legal relationship. In addition to the individual contractual relationships growing out of the traditional employment situation, labour law deals with the statutory requirements and collective relationships that are increasingly important in mass-production societies, the legal relationships between organized economic interests and the state, and the various rights and obligations related to some types of social services.

Labour law has won recognition as a distinctive branch of the law within the academic legal community, but the extent to which it is recognized as a separate branch of legal practice varies widely depending partly on the extent to which there is a labour code or other distinctive body of labour legislation in the country concerned, partly on the extent to which there are separate labour courts or tribunals, and partly on the extent to which an influential group within the legal profession practice specifically as labour lawyers.

In the early phases of development the scope of labour law is often limited to the most developed and important industries, to undertakings above a certain size, and to wage earners; as a general rule, these limitations are gradually eliminated and the scope of the law extended to include handicrafts, rural industries and agriculture, small undertakings, office workers, and, in some countries, public employees. Thus a body of law originally intended for the protection of manual workers in industrial enterprises is gradually transformed into a broader body of legal principles and standards, which have basically two functions: the protection of the worker as the weaker party in the employment relationship, and the regulation of the relations between organized interest groups (industrial relations).

Statute versus contract

The general tendency in the modern development of labour law has been the strengthening of statutory requirements and collective contractual relations at the expense of rights and obligations created by individual employment relationships. How important these latter remain depends, of course, on the degree of personal freedom in the given society as well as the autonomy of both employer and worker allowed by the actual operation of the economy. In such matters as hours of work, health and safety conditions, or industrial relations, the statutory or collective elements may define most of the substance of the rights and obligations of the individual worker, while with respect to such things as the duration of his appointment, his level and extent of responsibility, or his place in the scale of remuneration, these elements may provide what is essentially a framework for individual agreement.

HISTORICAL DEVELOPMENT OF LABOUR LAW

Impact of the Industrial Revolution

The origins of labour law can be traced back to the remote past and the most varied parts of the world. While European writers often attach importance to the guilds and apprenticeship systems of the medieval world, some Asian scholars have identified labour standards as far back as the Laws of Hammurabi and rules for labour–management relations in the Laws of Manu; Latin-American authors point to the Laws of the Indies promulgated by Spain in the 17th century for its New World territories. None of these can be regarded as more than anticipations, with only limited influence on subsequent developments. Labour law as it is known today is essentially the child of successive industrial revolutions from the 18th century onward. It became necessary when customary restraints and the intimacy of employment relationships in small communities ceased to provide adequate protection against the abuses incidental to new forms of mining and manufacture on a rapidly increasing scale at precisely the time when the 18th-century Enlightenment, the French Revolution, and the political forces that they set in motion were creating the elements of the modern social conscience. It developed rather slowly, chiefly in the more industrialized countries of western Europe, during the 19th century and has attained its present importance, relative maturity, and worldwide acceptance only during the 20th century.

The first landmark of modern labour law was the British Health and Morals of Apprentices Act of 1802, sponsored by the elder Sir Robert Peel. Similar legislation for the protection of the young was adopted in Zürich in 1815 and in France in 1841. By 1848 the first legal limitation of the working hours of adults was adopted by the Landsgemeinde (citizens' assembly) of the Swiss canton of Glarus. Sickness insurance and workmen's compensation were pioneered by Germany in 1883 and 1884, and compulsory arbitration in industrial disputes was introduced in New Zealand in the 1890s. The progress of labour legislation outside western Europe, Australia, and New Zealand was slow until after World War I. The more industrialized states of the United States began to enact such legislation toward the end of the 19th century, but the bulk of the present labour legislation of the United States was not adopted until after the Depression of the 1930s. There was virtually no labour legislation in Russia prior to the October Revolution of 1917. In India children between the ages of seven and 12 were limited to nine hours of work per day in 1881 and adult males in textile mills to 10 hours per day in 1911, but the first major advance was the amendment of the Factory Act in 1922 to give effect to conventions adopted at the first session of the International Labour Conference at Washington, D.C., in 1919. In Japan rudimentary regulations on work in mines were introduced in 1890, but a proposed factory act was controversial for 30 years before it was adopted in 1911, and the decisive step was the revision of this act in 1923 to give effect to the Washington Convention on hours of work in industry. Labour legislation in Latin America began in Argentina in the early years of the century and received a powerful impetus from the Mexican Revolution, which ended in 1917, but, as in North America, the trend became general only with the impact of the Great Depression. In Africa the progress of labour legislation became significant only from the 1940s onward.

The legal recognition of the right of association for trade union purposes has a distinctive history. There is no other aspect of labour law in which successive phases of progress and regression have been more decisively influenced by political changes and considerations. The legal prohibition of such association was repealed in the United Kingdom in 1824 and in France in 1884; there have been many subsequent changes in the law and may well be further changes, but these have related to matters of detail rather than to fundamental principles. In the United States freedom of association for trade union purposes remained precarious and subject to the unpredictable scope of the labour injunction, by means of which the courts helped restrain trade union activity until the 1930s. The breakthrough for trade unionism and collective bargaining was achieved by the National Labor Relations Act of 1935. In many other countries the record of progress and regression with respect to freedom of association falls into clearly distinguished periods separated by decisive political changes. This has certainly been the case with Germany, Italy, Spain, Japan, and much of eastern Europe; there have been many illustrations of it, and there may well be more in the developing world.

Labour codes or other forms of comprehensive labour legislation and ministries of labour were not introduced until the 20th century. The first labour code (which, like many of its successors, was a consolidation rather than a codification) was projected in France in 1901 and promulgated in stages from 1910 to 1927. Among the more advanced formulations affecting the general condition of labour were the Mexican Constitution of 1917 and the Weimar Constitution of Germany of 1919, both of which gave constitutional status to certain general principles of social policy regarding economic rights. Provisions of this kind have become increasingly common and are now widespread in all parts of the world.

Departments or ministries of labour responsible for the effective administration of labour legislation and for promoting its future development were established in Canada in 1900, in France in 1906, in the United States in 1913, in the United Kingdom in 1916, and in Germany in 1918. They became general in Europe and were established in India and Japan during the following years and became common in Latin America in the '30s. A labour office was established in Egypt in 1930, but only in the '40s and '50s did similar arrangements begin to take root elsewhere in Asia and Africa. Under differing political circumstances there continue, of course, to be wide variations in the authority and effectiveness of such administrative machinery.

ELEMENTS OF LABOUR LAW

The basic subject matter of labour law can be considered under nine broad heads: employment; individual employment relationships; wages and remuneration; conditions of work; health, safety, and welfare; social security; trade unions and industrial relations; the administration of labour law; and special provisions for particular occupational or other groups.

Employment. Employment considered as a basic concept and category of labour law is a relatively recent development. Prior to the Great Depression and World War II the emphasis was upon the prevention or reduction of excessive unemployment rather than upon long-term employment policy as part of a comprehensive scheme to promote economic stability and growth. The new approach, arising from changes in political outlook and contemporary economic thought, has increasingly found expression in legal provisions that establish the creation of employment opportunities as a general objective of policy. To this end, legislation has established the necessary legal framework for the forecasting of manpower needs and availability and the provision of employment services including placement, recruitment, vocational training, and apprenticeship. Freedom from forced labour, equality of treatment in employment and occupation, and unemployment benefits may, in a broad sense, be regarded as part of the same general subject.

The worker in new economic thought

Individual employment relations. The making, modification, and termination of individual employment relations and the resulting obligations for the parties form a second branch of labour law. It may also involve certain aspects of promotion, transfer, and dismissal procedures and compensation. Historically speaking, the law on these matters was at one time described as the law of master and servant. It implied a contractual relation in which one party agreed to be under the control of the other in the sense that the servant was bound to obey orders not only as to the work that he would execute but also as to the details of the work and the manner of its execution. In return, the master had to pay a wage and grant certain minimum conditions for the protection of the worker. As the law developed, the implied terms and statutory incidents attached to this relationship concerning such matters as termination of employment, dismissal procedures and compensation, minimum wages, conditions of work, and social security rights began to limit freedom of contract. The individual employment relationship continues, however, to be the subject matter of labour law to which general legal principles, as opposed to statutes and collective agreements, apply. Legally speaking, the individual contract of employment plays a more important role in the civil-law countries than in common-law countries.

Wages and remuneration. The substantive law on wages and remuneration covers such elements as forms and methods of payment, the protection of wages against unlawful deductions and other abuses, minimum wage arrangements, the determination of wages, fringe benefits, and, in highly sophisticated economies, incomes policies. The concept of wage regulation as a restraint upon extreme social evils has gradually been superseded by wage policies as deliberate instruments of positive management designed to promote economic stability and growth.

Wage protection and abuse

Legal requirements concerning the forms of wages and methods of wage payment deal with such matters as the proper notification of wage conditions, the payment of wages in legal tender or by check, the limitation and proper valuation of payments in kind, the freedom of the worker to dispose of his wages, regularity in wage payments, the treatment of wages as a privileged, or secured, debt, and restrictions upon the attachment or assignment of wages.

Minimum-wage regulation takes varied forms; it may, following the pattern originally set by the British Trades Boards Acts from 1909 onward, provide for wages councils or similar bodies to fix wages in trades that have no arrangements for collective agreements and where wages are exceptionally low; it may consist, as in Australia and New Zealand, essentially of arbitration arrangements; or it may, as in the United States under the Fair Labor Standards acts, provide a statutory rate or criteria for determining such a rate. Statutory provisions and collective agreements for determining wages may embrace such varied matters as skill differentials, the elimination of race and sex differentials, payment according to results and the relationship of wages to productivity, and wage guarantees for agreed periods of time. Fringe benefits, such as bonuses payable in varying contingencies, are typically a matter for collective agreements. Incomes policies remain the subject of much controversy. Their general purpose, sometimes embodied in legislation and sometimes expressed in collective agreements or statements of government policy, is to restrain inflationary pressures resulting from wage increases unrelated to increased productivity and to do this in a manner that promotes a fairer distribution of income.

Conditions of work. The conditions of work involve hours, rest periods, and vacations; the prohibition of child labour and regulation of the employment of young persons; and special provisions concerning the employment of women. This part of the law originated in legislation for the protection of children, young persons, and women against the worst evils of the Industrial Revolution. It originally dealt particularly with such matters as admission to employment, night work, and excessive hours, but the elements of its content and their relative importance have been wholly transformed during the 20th century.

As economic and educational progress and changed social habits have limited child labour in the industrialized countries, and increasingly in the modernized sectors of developing economies, the special concern of labour law with regard to the young has shifted to such areas as vocational guidance and training, career planning and advancement, and medical protection.

As employment opportunities for women have become more varied and responsible, there has been a similar shift of emphasis from protective legislation, which has come to be regarded as discriminatory since it tends to limit such opportunities, to legal guarantees of equal pay and equal employment, coupled with adequate maternity protection and the provision of facilities to enable women with family responsibilities to continue to be employed.

Whereas previously any statutory limitation of the hours of work of adult males was regarded as being highly questionable, except in mines where it had been introduced on safety grounds, in a society of much increased leisure it has now become a general practice to fix maximum hours of work by statute or collective agreement. In many countries the eight-hour day has been superseded by the 40-hour week as the statutory maximum for a wide range of occupations, and collective agreements providing for substantially shorter working hours are not uncommon. The details of hours regulation, whether by statute or collective agreement, include such matters as exceptions and adjustments necessary for continuous shift working. In addition, such regulations cover the extensions permitted for preparatory, complementary, and intermittent work; the special rules for force majeure (work of absolute necessity), accident, maintenance, and repair work; and the limitation, authorization, and remuneration of overtime.

The principle of resting one day of the week, sanctioned as it is by religious practice in many places, was widely incorporated in legislation at an early date; the lengthening of this weekly rest through the creation of the five-day week has been strongly influenced by statutory requirements and collective agreements.

Holidays

Legislation granting annual holidays with pay and collective agreements providing for such holidays are almost entirely a development of the mid-20th century but are increasingly common; moreover, there is a marked tendency for the minimum annual holiday to be increased.

Complex questions may arise concerning the qualifying period of service required for entitlement, breaks in the continuity of service, the calculation of average or normal remuneration for the purpose of the holidays, the extent to which holidays may be divided, and the liability for holidays where there has been a change of employer.

Health, safety, and welfare. Such general matters as occupational health and accident prevention regulations and services; special regulations for hazardous occupations such as mining, construction, and dock work; and provisions concerning such health and safety risks as poisons, dangerous machinery, dust, noise, vibration, and radiation constitute the health, safety, and welfare category of labour law. The efforts of organized safety movements and the progress of occupational medicine have produced comprehensive occupational health and accident-prevention services and regulations no longer limited to a few specially acute risks but covering the full range of dangers arising from modern industrial processes. Major developments include increased concern with the widespread use of chemicals and increasing provision for welfare facilities related to employment, including feeding, rest, recreation, and transport facilities.

Social security. Social security ranges from basic employers' liability for occupational accidents to comprehensive schemes that include income security in the form of sickness, unemployment, retirement, employment injury, maternity, family, invalidity, and survivors' benefits and medical care. As with other aspects of labour law, a progression from the particular to the general has been characteristic of the development of social security legislation. By the time of World War I, workmen's compensation schemes were general in industrialized and industrializing countries, but they were highly restrictive in their provisions for specific cases. Pension insurance was part of Otto von Bismarck's legacy to Germany, but elsewhere there was little more to be found than pension funds for the privileged or noncontributory pensions for the aged. Great Britain had been the pioneer in health and unemployment insurance. But social insurance remained a pragmatic experiment limited to a few countries advanced in both economic development and social policies. The coverage was limited to specific risks for certain categories of protected persons. Its object was to protect the worker against the hazards of life for which preindustrial societies provide by some form of community or family responsibility, but the approach was piecemeal and was limited to the most manageable cases of acute hardship. Eventually, the impact of the world economic depression of the 1930s and World War II in the industrial nations and the increasingly apparent inadequacy of earlier forms of community responsibility in developing countries transformed the position. The concept of social security, first given statutory expression in the United States in 1935 and in New Zealand in 1938, superseded that of social insurance, and the Beveridge Report of 1943 developed it even further to provide a basic income for all in need of such protection, in addition to providing comprehensive medical care. The concept has continued to broaden since that time, and social security has found increasing acceptance, though necessarily with varying degrees of practical application, in countries in the most varied stages of economic development.

The right to security

Acute, sometimes highly controversial, problems, particularly in the cost and efficiency of administrative organization of social security programs and of medical care, remain almost everywhere. But many countries have made progress in making higher standards of medical care available as a legal right and in converting the guarantee of a basic income as a protection against want into provision for effective income maintenance in the event of unemployment or loss of the family breadwinner. The idea is still developing. The trend is to broaden it to the point at which it includes all the varied hazards of life, including accidents of any kind, with the idea of facilitating economic growth by reducing the human cost of structural change. The pattern varies widely in different countries, partly as a reflection of different relationships between social security and private life, retirement, and health insurance, and partly because of differences in economic and social conditions.

Trade unions and industrial relations. A number of complex legal relationships fall under the heading of industrial relations, including the legal status, rights, and obligations of trade unions and employers' organizations, collective bargaining and collective agreements, the representation of employees at plant and enterprise level (including joint consultation and, where it exists, codetermination and other forms of workers' participation in management, even to the extent of workers' representation on company boards), and the prevention and settlement of various types of labour disputes in general and of strikes and lockouts in particular. There are wide variations both in the extent to which such matters as the representative character and capacity of trade unions, their legal status, the obligation to recognize and bargain with them, the enforceability of collective agreements, the scope of activities permitted to trade unions, and their obligations in contract and tort are subject to legal rules and in the content of such rules. In the United States, for instance, there is a considerable body of law on these subjects, the most important enactments being the National Labor Relations Act of 1935 (the Wagner Act) and the Labor Management Relations Act of 1947 (the Taft–Hartley Act); in the United Kingdom the law has hitherto remained marginal to most of these trade union issues, except for legislation of 1871, 1875, and 1906, which had provided certain "immunities," or "privileges," as they are sometimes called, for trade unions, particularly in connection with trade disputes. Legislation enacted in the early 1980s restricted some of these immunities or privileges, the trend being to expand the role of law in labour–management relations to reduce the increasing disruption caused by industrial conflict in a complex society. How to reconcile freedom of association and collective bargaining with the stability and growth of the economy remains the most challenging and difficult problem of labour law.

The administration of labour law. Another feature of labour law involves the organization and functioning of administrative authorities such as labour departments, labour inspection services, and other organs of enforcement. Administration of the law also encompasses the operation of labour courts and other bodies for the settlement of grievances arising from existing contracts or collective agreements and of industrial disputes arising between labour and management.

Labour law and the economic goal

The principal problem in many countries is to relate the process of labour administration and its special intimacy with labour and management to overall economic and social planning in a manner that gives proper weight to social considerations in economic policy. This problem falls mostly outside the scope of labour law, but its solution does depend in part on the extent to which labour law provides for and secures effective standards of administration.

Special categories of workers. Labour law includes many provisions for particular occupational or other groups. These sometimes appear as special parts of a general code, special legislation, or provisions that limit specific legislative provisions with regard to particular groups. These special provisions are common and important in mining, transportation (and in particular maritime transport), commercial occupations, and agriculture. Cutting across these broad sectors of economic activity are the traditional legal distinctions made in some countries between blue-collar workers and salaried employees and certain newer distinctions, such as that between employees who earn annual salaries and have rights of tenure and persons with no such rights engaged and remunerated on a monthly, weekly, or even daily basis.

THE RULES OF DIFFERENT SYSTEMS

Among the distinctive elements of labour law that reflect the political, socioeconomic, and legal differences among countries are variations in the relative importance of statutory regulation and collective agreements, the prevalence of national or industrial collective agreements as opposed to company or plant agreements, the importance in certain countries of arbitral awards, and the extent to which labour law has been affected by a country's constitutional structure, especially with regard to judicial review of constitutionality of legislation and judicial interpretation of constitutional powers, limitations, and guarantees.

In the United Kingdom, for instance, the tradition has been to allow a maximum of initiative and freedom to employers' and workers' organizations in the regulation of their mutual relations and the determination of conditions of work. Most countries on the Continent, by contrast, have detailed legislative provisions on these matters.

In the United Kingdom, however, the reluctance to legislate is becoming less marked; there is now legislation concerning industrial training and discrimination in employment, formerly matters for collective agreement; and legislation concerning collective bargaining, safeguards against unfair dismissal, and certain trade union practices was enacted in the late 1970s and early 1980s. In virtually all of the developing countries the absence of an established tradition of collective bargaining and the importance of the part played by the state in economic development have placed a premium on legislative action.

Variations in collective agreements

The coverage and scope, term of validity, and legal effect of collective agreements vary widely. In Sweden there has been a practice of national negotiations covering the whole of industry; in the United Kingdom agreements generally cover an industry or occupation in the country as a whole or a particular industrial area; in the United States and in Japan the unit of negotiation is generally the company or plant. The contrast may be less significant in practice than in principle, since an important company or plant agreement tends to set an industry-wide pattern of negotiation (*e.g.,* in the automobile industry of the United States); nevertheless, the difference is important.

In Australia and New Zealand conciliation and arbitration tribunals determine by awards matters normally dealt with in other countries by legislation or collective agreement, such as wages, hours, and conditions of work. The example has had some influence on systems of arbitration courts established in developing countries, notably in Asia and East Africa, but with this exception there is no tendency for it to be widely imitated elsewhere except as a device for avoiding deadlocks in negotiation, especially in essential public services.

Constitutional impediments

In the United States and Canada the development of labour law has been affected by questions of constitutionality, which not only influenced its ultimate form but also retarded its development. In the United States the constitutionality of workmen's compensation laws was much debated until it was favourably settled by the Supreme Court in 1917; child-labour and minimum-wage regulations were delayed by judicial decisions holding them to be outside federal competence and, in some cases, inconsistent with the constitutional guarantee against deprivation by the state of life, liberty, and property without due process of law (the guarantee here applying to the factory owner). The first attempt of President Franklin Roosevelt's administration to regulate hours and wages by codes of fair competition during the Great Depression was also held to be unconstitutional as an improper delegation of legislative power by Congress to the executive branch. But thereafter the temper of judicial review changed and the validity of federal legislation guaranteeing free collective bargaining, regulating wages and hours, and establishing social security was upheld. In Canada, a pioneer in establishing a labour department, restrictive judicial interpretations of the powers of the federal government had a similar effect, and only after World War II did federal–provincial cooperation afford a basis for achieving greater uniformity and more rapid progress.

UNIFYING TENDENCIES

The range of possible solutions for similar problems often consists of variations of detail on a limited number of options, and certain common elements, often expressed in identical or almost identical texts, recur in the law of different countries. These elements derive partly from the legislation of other nations but increasingly from the influence on the law of the international standards evolved by the International Labour Organisation (ILO).

From the beginnings of modern labour legislation in the early 19th century, the law of certain countries has been extensively used by other countries as a model. For example, British factory legislation was widely copied at an early date and German social-insurance legislation provided a prototype from the time of Bismarck's reforms. British legislation has continued to serve as a model for the basic legislation of many states that were formerly British dependencies and remains in force subject to modifications made since independence. Much of the French Labour Code became applicable through the 1952 Labour Code for Overseas Territories to the states that were formerly French dependencies and remains the basis of their labour law. The United States legislation of the period from the 1930s onward has been exported to Japan, the Philippines, Liberia, and other countries. The Mexican Labour Law of 1931, varied by elements derived primarily from European models, had considerable influence on the early development of labour law in a number of Latin-American countries. Through quite another process, the labour law of the Soviet Union has reshaped without replacing some of the earlier elements in the labour law of the other Socialist states. But on the whole these national influences of particular countries and legal systems are declining.

The ILO: international standards

During the mid-20th century the standards evolved by the ILO became the leading external influence upon the labour law of many countries. They have had a far-reaching impact in virtually all of the advanced countries except the United States and the Soviet Union, where external influences have been secondary. In much of the developing world they were of great importance even before independence, since much of the legislation sponsored there by the colonial powers was based on ILO standards.

The ILO, created in 1919 as an autonomous partner of the League of Nations, and since 1946 a specialized agency associated with the United Nations, adopts international standards in the form of conventions and recommendations. Conventions when ratified become binding obligations of the member states ratifying them; recommendations are designed as guides for legislation, collective agreements, administrative measures, etc. Elaborate follow-up arrangements, including examination of regular reports and commissions of enquiry into complaints, are provided to ensure that the obligations assumed are fulfilled. These standards, which already cover, in varying degrees of detail and at varied stages of development, virtually all the more important branches of labour law, are constantly amplified and revised at the annual sessions of the International Labour Conference.

Unification or, as the process is often called, harmonization of labour law is one of the professed purposes of a number of regional organizations in different parts of the world, but only in the Council of Europe, the European Communities, and the Organization of American States have tangible measures been taken, largely based on ILO standards. The more important instruments adopted are the European Social Charter, the European Social Security Code, the Social Security Regulations of the European Economic Community, and the Central American Convention on Social Security for Migrant Workers. The Conferences of American and African Labour Ministers, sponsored by the Organization of American States and the Organization of African Unity respectively, and the Conference of Asian Labour Ministers, which has developed without any comparable sponsorship, discuss matters of general policy and the coordination of action in the ILO rather than the formulation of specific standards. An Arab Labour Organization was created in 1970.

CONTEMPORARY TENDENCIES

Labour law differs from the older branches of the law in that its history has been in some cases so much influenced by the ebb and flow of political change, its development so rapid, and its expansion on a world scale so recent, that it is difficult to predict its future. But the trend is clear. In no place is labour law losing importance. While some types of protective legislation, notably special provision for the protection of women workers, are losing their importance, the tendency is toward more comprehensive legislation embracing a wider range of subjects and often dealing with matters previously left to collective agreement, individual contract, or the discretion of the employer.

The transition everywhere has been from a class law protecting the weakest segment of society to a community law designed to serve the common interest. This development is seen in the elimination of limitations and exceptions to the law and in the increasing emphasis given to matters of general interest, including full employment, equitable distribution of wealth, and community responsibility for the incidence of misfortune in individual lives.

The place of labour law in legal evolution

Labour law must also be said to serve the social interest in promoting constructive industrial relations and reducing the occurrence of open conflict. This evolution of labour law is an important contribution to the evolution of the law as a whole, from a law for the propertied and trading classes with a special chapter for the working class to a common law for the entire community.

The importance of a body of law that has a dynamic and progressive impact rather than a restrictive influence is now widely understood, and the need for legal flexibility to facilitate economic and social development and change is increasingly appreciated. In addition, the value of delegated powers and procedures of consultation with interested groups and organizations to achieve such flexibility is more generally recognized. Social objectives remain the test of the validity of economic policy, and labour law plays a major part in defining these objectives and ensuring that economic policy respects them in the interest of the whole community. (C.W.J./J.Sch./Ed.)

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Business Organization

In the widest sense, all social groupings are organized for some form of business. This article, however, is restricted to the contemporary concept of an organization formed for the purpose of carrying on commercial enterprise. Such a concept is predicated on systems of law governing contract and exchange, property rights, and incorporation. In this narrower sense, business enterprises customarily take one of three forms: individual proprietorships, partnerships, or limited-liability companies (or corporations). In the first form, a single person holds the entire operation as his personal property, usually managing it on a day-to-day basis. Most businesses are of this type. Of the nearly 15,000,000 business establishments in the United States in the late 20th century, for example, more than three-quarters were individual proprietorships; they tended to be small operations, most of them with annual receipts of less than $50,000. The second form, the partnership, may have from two to 50 or more members, as in the case of large law and accounting firms, brokerage houses, and advertising agencies. This form of business is owned by the partners themselves; they may receive varying shares of the profits depending on their investment or contribution. Whenever a member leaves or a new member is added, the firm must be reconstituted as a new partnership. The third form, the limited-liability company, or corporation, denotes incorporated groups of persons—that is, a number of persons considered as a legal entity (or fictive "person") with property, powers, and liabilities separate from those of its members. This type of company is also legally separate from the individuals who work for it, whether they be shareholders or employees or both; it can enter into legal relations with them, make contracts with them, and sue and be sued by them. Most large industrial and commercial organizations are limited-liability companies, or corporations.

This article deals primarily with the large private business organizations made up chiefly of partnerships and limited-liability companies—called collectively business associations. Some of the principles of operation included here also apply to large individually owned companies and to public enterprises.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 553, and the *Index.*

The article is divided into the following sections:

Types of business associations

Business associations have three distinct characteristics: (1) they have more than one member (at least when they are formed); (2) they have assets that are legally distinct from the private assets of the members; and (3) they have a formal system of management, which may or may not include members of the association.

Essential features of business associations

The first feature, plurality of membership, distinguishes the business association from the business owned by one individual; the latter does not need to be regulated internally by law because the single owner totally controls the assets. Because the single owner is personally liable for debts and obligations incurred in connection with the business, no special rules are needed to protect its creditors beyond the ordinary provisions of bankruptcy law.

The second feature, the possession of distinct assets (or a distinct patrimony), is required for two purposes: (1) to delimit the assets to which creditors of the association can resort to satisfy their claims (though in the case of some associations, such as the partnership, they can also compel the members to make good any deficiency) and (2) to make clear what assets the managers of the association may use to carry on business. The assets of an association are contributed directly or indirectly by its members—directly if a member transfers a business or property or investments of his own to the association in return for a share in its capital, and indirectly if a member pays his share of capital in cash and the association then uses his contribution and like contributions in cash made by other members to purchase a business, property, or investments.

Limited participation of shareholders

The third essential feature, a system of management, varies greatly. In a simple form of business association the members who provide the assets are entitled to participate in the management unless otherwise agreed. In the more complex form of association, such as the company or corporation of the Anglo-American common-law countries, members have no immediate right to participate in the management of the association's affairs; they are, however, legally entitled to appoint and dismiss the managers (known also as directors, presidents, or administrators), and their consent is legally required (if only pro forma) for major changes in the company's structure or activities, such as reorganizations of its capital and mergers with other associations. The role of a member of a company or corporation is basically passive; he is known as a shareholder or stockholder, the emphasis being placed on his investment function. The managers of a business association, however, do not in law comprise all of the persons who exercise discretion or make decisions. Even the senior executives of large corporations or companies may be merely employees, and, like manual or clerical workers, their legal relationship with the corporation is of no significance in considering the law governing the corporation. Whether an executive is a director, president, or administrator (an element in the company or corporation's legal structure) depends on purely formal considerations; if he is named as such in the document constituting the corporation, or if he is subsequently appointed or elected to hold such an office, it is irrelevant whether his actual functions in running the corporation's business and the power or influence he wields are great or small. Nevertheless, for certain purposes, such as liability for defrauding creditors in English law and liability for deficiencies of assets in bankruptcy in French law, people who act as directors and participate in the management of the company's affairs are treated as such even though they have not been formally appointed.

PARTNERSHIPS

The
imited
artnership

The distinguishing features of the partnership are the personal and unrestricted liability of each partner for the debts and obligations of the firm (whether he assented to their being incurred or not) and the right of each partner to participate in the management of the firm and to act as an agent of it in entering into legal transactions on its behalf. The civil-law systems of most continental European countries have additionally always permitted a modified form of partnership, the limited partnership (*société en commandite, Kommanditgesellschaft, società in accomandita*) in which one or more of the partners are liable for the firm's debts only to the extent of the capital they contribute or agree to contribute. Such limited partners are prohibited from taking part in the management of the firm, however; if they do, they become personally liable without limit for the debts of the firm, together with the general partners. English common law refused to recognize the limited partnership, and in the United States at the beginning of the 19th century only Louisiana, which was governed by French civil law, permitted such partnerships. During the 19th century most of the states enacted legislation allowing limited partnerships to be formed, and in 1907 Great Britain adopted the limited partnership by statute, but it has not been much used there in practice. Another distinction between kinds of partnership in civil law—one that has no equivalent in Anglo-American common-law countries—is that between civil and commercial partnerships. This distinction depends on whether or not the purposes for which the partnership is formed fall within the list of commercial activities in the country's commercial code. These codes always make manufacturing, dealing in, and transporting goods commercial activities, while professional and agricultural activities are always noncommercial. Consequently, a partnership of lawyers, doctors, or farmers is a civil partnership, governed exclusively by the civil code of the country concerned and untouched by its commercial code. No such distinction is made in the common-law countries, where professional and business partnerships are subject to the same rules as trading partnerships, although only partners in a trading partnership have the power to borrow on the firm's behalf.

LIMITED-LIABILITY COMPANIES, OR CORPORATIONS

The company or corporation, unlike the partnership, is formed not simply by an agreement entered into between its first members; it must also be registered at a public office or court designated by law, or otherwise obtain official acknowledgment of its existence. Under English and American law the company or corporation is incorporated by filing the company's constitution (memorandum and articles of association, articles or certificate of incorporation) signed by its first members at the Companies Registry in London or, in the United States, at the office of the state secretary of state or corporation commissioner. In France, West Germany, and Italy and the other countries subject to a civil-law system, a notarized copy of the constitution is filed at the local commercial tribunal, and proof is tendered that the first members of the company have subscribed the whole or a prescribed fraction of the company's capital, and that assets transferred to the company in return for an allotment of its shares have been officially valued and found to be worth at least the amount of capital allotted for them. English and American law, together with the laws of The Netherlands and the Scandinavian countries, provide only one category of business company or corporation (in The Netherlands the *naamloze vennootschap,* in Sweden the *aktiebolag*), although all these systems of law make distinctions for tax purposes between private, or close, companies or corporations on the one hand and public companies or corporations on the other. English law also distinguishes between private and public companies for some purposes of company law; for example, a private company cannot have more than 50 members and cannot advertise subscriptions for its shares. Under the civil-law systems, however, a fundamental distinction is drawn between the public company (*société anonyme; Aktiengesellschaft; società per azioni*) and the private company (*société à responsabilité limitée, Gesellschaft mit beschränkter Haftung* [*G.m.b.H.*], *società a responsabilità limitata*), and in West Germany the two kinds of company are governed by different enactments, as they were in France until 1966. For practical purposes, however, public and private companies function the same way in all countries. Private companies are formed when there is no need to appeal to the public to subscribe for the company's shares or to lend money to it, and often they are little more than incorporated partnerships whose directors hold all or most of the company's shares. Public companies are formed—or more usually created by the conversion of private companies into public ones—when the necessary capital cannot be supplied by the directors or their associates and it is necessary to raise funds from the public by publishing a prospectus. In Great Britain, the Commonwealth countries, and the United States, this also requires the obtaining of a stock exchange listing for the shares or other securities offered or an offer on the Unlisted Securities Market (USM). In a typical public company the directors hold only a small fraction of its shares, often less than 1 percent, and in Great Britain and the United States, at least, it is not uncommon for up to one-half of the funds raised by the company to be represented not by shares in the company but by loan securities such as debentures or bonds.

Public and
private
companies

In Anglo-American common-law countries, public and private companies account for most of the business associations formed, and partnerships are entered into typically only for professional activities. In European countries the partnership in both its forms is still widely used for commercial undertakings. In West Germany a popular form of association combines both the partnership and the company. This is the G.m.b.H. & Co., which is a limited partnership whose general partner (nominally liable without limit for the partnership's debts) is a private company and whose limited partners are the same persons as the shareholders of the company. The limited partners enjoy the benefit of limited liability for the partnership's debts, and by ensuring that most of the partnership's profits are paid to them as limited partners and not to them as shareholders in the private company, they largely avoid the incidence of corporation tax.

Shares and other securities. Under all systems of law a partner may assign his share or interest in a partnership to anyone he wishes unless the partnership agreement forbids this, but the assignment does not make the assignee a partner unless all of the other partners agree. If they do not, the assignee is merely entitled to receive the financial benefits attached to the share or interest without being able to take part in the management of the firm, but neither is he personally liable for the debts of the firm.

How
company
shares
differ from
those in a
partnership

The shares of a company are quite different. In the first place, they are freely transferable unless the company's constitution imposes restrictions on their transfer, or, in French and Belgian law, unless the company is a private one, in which case transfers require the consent of the holders of three-quarters of the company's issued shares. The constitution of an English private company must always restrict the transfer of its shares for the company to qualify as private. The restriction is usually that the directors may refuse to register a transfer for any of several reasons, or that the other shareholders shall have the right to buy the shares at a fair price when their holder wishes to sell. In American law similar restrictions may be imposed, but unreasonable restrictions are disallowed by the courts. According to French and West German law, the transfer of shares in public companies may be restricted only by being made subject to the consent of the board of directors or of the management board, but under French law if the directors do not find an alternative purchaser at a fair price within three months their consent is considered as given.

Limited liability. The second significant difference between share holding and partnership is that shares in a company do not expose the holder to unlimited liability in the way that a partner (other than a limited one) is held liable for the debts of his firm. Under all systems of law, except those of Belgium and some of the U.S. states, all shares must have a nominal value expressed in money terms, such as $10, £1, DM. 50, or Fr. 100, the latter two being the minimum permissible under West German and French law. A company may issue shares for a price greater than this nominal value (the excess being known as a share premium), but it generally cannot issue them for less. Any part of that nominal value and the share premium that has not so far been paid is the measure of the shareholder's maximum liability to contribute if the company becomes insolvent. If shares are issued without a nominal value (no par value shares), the subscription price is fixed by the directors and is the measure of the shareholder's maximum liability to contribute. Usually the subscription price of shares is paid to the company fairly soon after they are issued. The period for payment of all the installments is rarely more than a year in common-law countries, and it is not uncommon for the whole subscription price to be payable when the shares are issued. The actual subscription price is influenced by market considerations, such as the company's profit record and prospects, and by the market value of the company's existing shares. Although directors have a duty to obtain the best subscription price possible, they can offer new shares to existing shareholders at favourable prices, and those shareholders can benefit either by subscribing for the new shares or by selling their subscription rights to other persons. Under European legislation directors are bound to offer new shares to existing shareholders in the first place unless they explicitly forgo their preemptive rights. In most U.S. states (but not in the United Kingdom) such preemptive rights are implied if the new shares belong to the same class as existing shares, but the rights may be negated by the company's constitution.

Dividend declaration

Dividends. The third difference between share holding and partnerships is that a partner is automatically entitled to his share of profits of the firm as soon as they are ascertained, but a shareholder is only entitled to a dividend out of the company's profits when it has been declared. Under English law, dividends are usually declared at annual general meetings of shareholders, though the company's constitution usually provides that the shareholders cannot declare higher dividends than the directors recommend. Under American law dividends are usually declared by the directors, and if shareholders consider, in view of the company's profits, that too small a dividend has been paid, they may apply to the court to direct payment of a reasonable dividend. West German law similarly protects shareholders of public companies against niggardly dividends by giving the annual general meeting power to dispose as it wishes of at least half the profit shown by the company's annual accounts before making transfers to reserve. For the same object Swedish law empowers the holders of 10 percent of a company's shares to require at least one-fifth of its accumulated profits and reserves to be distributed as a dividend, provided that the total distribution does not exceed one-half of the profits of its last financial year. Thus, most national law recognizes potential conflict of interest between directors and shareholders.

Classes of shares. Companies may issue shares of different classes, the commonest classes being ordinary and preference, or in U.S. terminology, common and preferred shares. Preference shares are so called because they are entitled by the terms on which they are issued to payment of a dividend of a fixed amount (usually expressed as a percentage of their nominal value) before any dividend is paid to the ordinary shareholders. In the case of cumulative preference shares, any unpaid part of a year's dividend is carried forward and added to the next year's dividend, and so on until the arrears of preference dividend are paid off. The accumulation of arrears of preference dividend depreciates the value of the ordinary shares, whose holders cannot be paid a dividend until the arrears of preference dividend have been paid. Consequently, it has been common in the United States (but not in the United Kingdom) for companies to issue noncumulative preference shares, giving their holders the right to a fixed preferential dividend each year if the company's profits are sufficient to pay it, but limiting the dividend to the amount of the profits of the year if they are insufficient to pay the preference dividend in full. Preference shares are not common in Europe, but under West German and Italian law they have the distinction of being the only kind of shares that can be issued without voting rights in general meetings, all other shares carrying voting rights proportionate to their nominal value by law.

History of the limited-liability company. The limited-liability company, or corporation, is a relatively recent innovation. Only since the mid-19th century have incorporated businesses risen to ascendancy over other modes of ownership. Thus, any attempt to trace the forerunners of the modern corporation should be distinguished from a general history of business or a chronicle of associated activity. Men have embarked on enterprises for profit and have joined together for collective purposes since the dawn of recorded history, but these early enterprises were forerunners of the contemporary corporation in terms of their functions and activities, not in terms of their mode of incorporation. When a group of Athenian or Phoenician merchants pooled their savings to build or charter a trading vessel, their organization was not a corporation but a partnership; ancient societies did not have laws of incorporation that delimited the scope and standards of business activity.

Corporations in the Middle Ages

The corporate form itself developed in the early Middle Ages with the growth and codification of civil and canon law. Several centuries, however, passed before business ownership was subsumed under this arrangement. The first corporations were towns, universities, and ecclesiastical orders. These differed from partnerships in that the corporation existed independently of any particular membership. Unlike modern business corporations, they were not the "property" of their participants. The holdings of a monastery, for example, belonged to the order itself; no individual owned shares in its assets. The same was true of the medieval guilds, which dominated many trades and occupations. As corporate bodies, they were chartered by government, and their business practices were regulated by public statutes; each guild member, however, was an individual proprietor who ran his own establishment, and, while many guilds had substantial properties, these were the historic accruals of the associations themselves. By the 15th century, the courts of England had agreed on the principle of "limited liability": *Si quid universitati debetur, singulis non debetur, nec quod debet universitas, singuli debent* ("If something is owed to the group, it is not owed to the individuals nor do the individuals owe what the group owes"). Originally applied to guilds and municipalities, this principle set limits on how much an alderman of the Liverpool Corporation, for example, might be called upon to pay if the city ran into debt or bankruptcy. Applied later to stockholders in business corporations, it served to encourage investment because the most an individual could lose in the event of the firm's failure would be the actual amount he had originally paid for his shares.

Incorporation of business enterprises began in England during the Elizabethan era. This was a period when businessmen were beginning to accumulate substantial surpluses, and overseas exploration and trade presented expanded investment opportunities. This was an age that gave overriding regulatory powers to the state, which sought to ensure that business activity was consonant with current mercantilist conceptions of national prosperity. Thus, the first joint-stock companies, while financed with private capital, were created by public charters setting down in detail the activities in which the enterprises might operate. In 1600, Queen Elizabeth I granted to a group of investors headed by the Earl of Cumberland the right to be "one body corporate," known as the Governor and Company of Merchants of London, trading into the East Indies. The East India Company was bestowed a trading

The East India Company

monopoly in its territories and also was given authority to make and enforce laws in the areas it entered. The East India Company, the Royal African Company, the Hudson's Bay Company, and similar incorporated firms were semipublic enterprises acting both as arms of the state and as vehicles for private profit. The same principle held with the colonial charters on the American continent. In 1606 the crown vested in a syndicate of "loving and well-disposed Subjects" the right to develop Virginia as a royal domain, including the power to coin money and to maintain a military force. The same was done, in subsequent decades, for the "Governor and Company of the Massachusetts Bay in New England," and for William Penn's "Free Society of Traders" in Pennsylvania.

Much of North America's settlement was initially underwritten as a business venture. But, while British investors accepted the regulations inhering in their charters, American entrepreneurs came to regard such rules as repressive and unrealistic. The U.S. War of Independence can be interpreted as a movement against the tenets of this mercantile system, raising serious questions about a direct tie between business enterprise and public policy. One result of that war, therefore, was to establish the premise that a corporation need not show that its activities advance a specific public purpose. Alexander Hamilton, the first secretary of the treasury and an admirer of Adam Smith, took the view that businessmen should be encouraged to explore their own avenues of enterprise. "To cherish and stimulate the activity of the human mind, by multiplying the objects of enterprise, is not among the least considerable of the expedients by which the wealth of a nation may be promoted," he wrote in 1791.

19th-century statutes

The growth of independent corporations did not occur overnight. For a long time, both in Europe and in the United States, the corporate form was regarded as a creature of government, providing a form of monopoly. In the United States the new state legislatures granted charters principally to public-service companies intending to build or operate docks, bridges, turnpikes, canals, and waterworks, as well as to banks and insurance companies. Of the 335 companies receiving charters prior to 1800, only 13 were firms engaging in commerce or manufacturing. By 1811, however, New York had adopted a general act of incorporation, setting the precedent that businessmen had only to provide a summary description of their intentions for permission to launch an enterprise. By the 1840s and 1850s, the rest of the states had followed suit. In Great Britain after 1825, the statutes were gradually liberalized so that the former privilege of incorporating joint-stock companies became the right of any group complying with certain minimum conditions, and the principle of limited liability was extended to them. A similar development occurred in France and parts of what are now Germany.

By the 20th century, in terms of size, influence, and visibility, the corporation has become the dominant business form in Western industrial societies. While corporations may be large or small, ranging from firms having hundreds of thousands of employees to neighbourhood businesses of very modest proportions, public attention has increasingly focused on the several hundred giant companies that play a preponderant economic role in the United States, Japan, Korea, the nations of western Europe, Canada, Australia, New Zealand, South Africa, and several other countries. These firms not only occupy important positions in the economy, but they have great social, political, and cultural influence as well. Both at home and abroad they affect the operations of national and local governments, give shape to local communities, and influence the values of ordinary individuals. Therefore, while in fact and in law corporate businesses are private enterprises, their activities have consequences that are public in character and as pervasive as those of many governments.

OTHER FORMS OF BUSINESS ASSOCIATION

Besides the partnership and the company or corporation there are a number of other forms of business association, of which some are developments or adaptations of the partnership or company, some are based on contract between the members or on a trust created for their benefit, and others are statutory creations. The first of these classes includes the cooperative society; the building society or home loan association and its West German equivalent, the *Bausparkasse;* the trustee savings bank or people's or cooperative bank; the friendly society or mutual insurance association; and the American mutual fund investment company. The essential features of these associations are that they provide for the small or medium investor; and, although they originated as contractual associations, they are now governed in most countries by special legislation and not by the law applicable to companies or corporations.

Contractual associations

The establishment and management of cooperatives is treated in most countries under laws distinct from those governing other business associations. The cooperative is a legal entity but typically owned and controlled by those who use it or work in it, though there may be various degrees of participation and profit sharing. The essential point is that the directors and managers are accountable ultimately to the enterprise members, not to the outside owners of capital. This form is rooted in a strong sense of social purpose; it was devised more than a century ago as an idealistic alternative to the conventional capitalist business association. It has been particularly associated with credit, retailing, agricultural marketing, and crafts.

The second class comprises the English unit trust and the European *fonds d'investissements* or*Investmentfonds,* which fulfill the same functions as the U.S. mutual funds; the Massachusetts business trust (now little used but providing a means of limiting the liability of participants in a business activity like the limited partnership); the foundation (*fondation, Stiftung*), a European organization that has social or charitable objects and often carries on a business whose profits are devoted to those objects; and, finally, the cartel or trade association, which regulates the business activities of its individual members and is itself extensively regulated by antitrust and antimonopoly legislation.

The third class of associations, those wholly created by statute, comprise corporations formed to carry on nationalized business undertakings (such as the Bank of England and the German Federal Railways) or to coexist with other businesses in the same field (such as the Italian Istituto per la Ricostruzione Industriale) or to fulfill a particular governmental function (such as the Tennessee Valley Authority). Such statutory associations usually have no share capital, though they may raise loans from the public. They are regarded in European law as being creatures of public law, like departments and agencies of the government. In recent years, however, a hybrid between the state corporation and the privately owned corporation or company has appeared in the form of the mixed company or corporation (*société mixte*). In this kind of organization part of the association's share capital is held by the state or a state agency and part by private persons, this situation often resulting from a partial acquisition of the association's shares by the state. In only France and Italy are there special rules governing such associations; in the United Kingdom and West Germany they are subject to the ordinary rules of company law.

Management and control of companies

The structure of management

The simplest form of management is the partnership. In Anglo-American common-law and European civil-law countries, every partner is entitled to take part in the management of the firm's business, unless he is a limited partner; however, a partnership agreement may provide that an ordinary partner shall not participate in management, in which case he is a dormant partner but is still personally liable for the debts and obligations incurred by the other managing partners.

The management structure of companies or corporations is more complex. The simplest is that envisaged by English, Belgian, Italian, and Scandinavian law, by which the shareholders of the company periodically elect a board of directors who collectively manage the company's affairs and reach decisions by a majority vote but also have the right to delegate any of their powers, or even the whole management of the company's business, to one or more of their number. Under this regime it is common for a managing director (*directeur général, direttore generale*) to be appointed, often with one or more assistant managing directors, and for the board of directors to authorize them to enter into all transactions needed for carrying on the company's business, subject only to the general supervision of the board and to its approval of particularly important measures, such as issuing shares or bonds or borrowing. The U.S. system is a development of this basic pattern. By the laws of most states it is obligatory for the board of directors elected periodically by the shareholders to appoint certain executive officers, such as the president, vice president, treasurer, and secretary. The latter two have no management powers and fulfill the administrative functions that in an English company are the concern of its secretary; but the president and in his absence the vice president have by law or by delegation from the board of directors the same full powers of day-to-day management as are exercised in practice by an English managing director.

The most complex management structures are those provided for public companies under West German and French law. The management of private companies under these systems is confided to one or more managers (*gérants, Geschäftsführer*) who have the same powers as managing directors. In the case of public companies, however, West German law imposes a two-tier structure, the lower tier consisting of a supervisory committee (*Aufsichtsrat*) whose members are elected periodically by the shareholders and the employees of the company in the proportion of two-thirds shareholder representatives and one-third employee representatives (except in the case of mining and steel companies where shareholders and employees are equally represented), and the upper tier consisting of a management board (*Vorstand*) comprising one or more persons appointed by the supervisory committee but not from its own number. The affairs of the company are managed by the management board, subject to the supervision of the supervisory committee, to which it must report periodically and which can at any time require information or explanations. The supervisory committee is forbidden to undertake the management of the company itself, but the company's constitution may require its approval for particular transactions, such as borrowing or the establishment of branches overseas, and by law it is the supervisory committee that fixes the remuneration of the managers and has power to dismiss them.

The German two-tiered public company

The French management structure for public companies offers two alternatives. Unless the company's constitution otherwise provides, the shareholders periodically elect a board of directors (*conseil d'administration*), which "is vested with the widest powers to act on behalf of the company" but which is also required to elect a president from its members who "undertakes on his own responsibility the general management of the company," so that in fact the board of directors' functions are reduced to supervising him. The similarity to the West German pattern is obvious, and French legislation carries this further by openly permitting public companies to establish a supervisory committee (*conseil de surveillance*) and a management board (*directoire*) like the West German equivalents as an alternative to the board of directors–president structure.

Dutch and Italian public companies tend to follow the West German pattern of management, although it is not expressly sanctioned by the law of those countries. The Dutch *commissarissen* and the Italian *sindaci*, appointed by the shareholders, have taken over the task of supervising the directors and reporting on the wisdom and efficiency of their management to the shareholders.

SEPARATION OF OWNERSHIP AND CONTROL

Dispersion of ownership

The investing public is a major source of funds for new or expanding operations. As companies have grown, their need for funds has grown, with the consequence that legal ownership of companies has become widely dispersed. For example, in large U.S. corporations, shareholders may run into the hundreds of thousands and even more. Although large blocks of shares may be held by wealthy individuals or institutions, the total amount of stock in these companies is so large that even a very wealthy person is not likely to own more than a small fraction of it.

The chief effect of this stock dispersion has been to give effective control of the companies to their salaried managers. Although each company holds an annual meeting open to all stockholders, who may vote on company policy, these gatherings, in fact, tend to ratify ongoing policy. Even if sharp questions are asked, the presiding officers almost invariably hold enough proxies to override outside proposals. The only real recourse for a dissatisfied shareholder is to sell his stock and invest in a firm whose policies are more to his liking. (If enough shareholders do this, of course, the price of the stock will fall quite markedly, perhaps impelling changes in management personnel or company policy.) Occasionally, there are "proxy battles," when attempts are made to persuade a majority of shareholders to vote against a firm's managers (or to secure representation of a minority bloc on the board), but such struggles seldom involve the largest companies. It is in the managers' interest to keep the stockholders happy, for, if the company's shares are regarded as a good buy, then it is easy to raise additional capital through a new stock issue.

Thus, if a company is performing well in terms of sales and earnings, its executives will have a relatively free hand. If a company gets into trouble, its usual course is to agree to be merged into another incorporated company or to borrow money. In the latter case, the lending institution may insist on a new chief executive of its own choosing. If a company undergoes bankruptcy and receivership, the court may appoint someone to head the operation. But managerial autonomy is the rule. The salaried executives typically have the discretion and authority to decide what products and services they will put on the market, where they will locate plants and offices, how they will deal with employees, and whether and in what directions they will expand their spheres of operation.

EXECUTIVE MANAGEMENT

The markets that corporations serve reflect the great variety of humanity and human wants; accordingly, firms that serve different markets exhibit great differences in technology, structure, beliefs, and practice. The essence of competition and innovation lies in differentiation and change, so corporations are in general under degrees of competitive pressure to modify or change their existing offerings, to introduce new products or services. Similarly, as markets decline or become less profitable, they are under pressure to invent or discover new wants and markets. Resisting this pressure for change and variety are benefits derived from regulated manufacturing, from standardization of machines and tools, and from labour specialization. Every firm has to arrive at a mode of balancing change and stability, a conflict often expressed in distinctions drawn between capital and revenue and long- and short-term operations and strategy. Many corporations have achieved relatively stable product–market relationships, providing further opportunity for growth within particular markets and expansion into new areas. Such relative market control endows corporate executives and officers with considerable discretion over resources and, in turn, with considerable corporate powers. In theory these men and women are hired to manage someone else's property; in practice, however, many management officers have come increasingly to regard the stockholders as simply one of several constituencies to which they must report at periodic intervals through the year.

Management motivation

Managerial decision making. The guidelines governing management decisions cannot be reduced to a simple formula. Traditionally, economists have assumed that the goal of a business enterprise was to maximize its profits. There are, however, problems of interpretation with this simple assertion. First, over time the notion of "profit" is itself unclear in operational terms. Today's profits can be increased at the expense of profits years away, by cutting maintenance, deferring investment, and exploiting staff. Second, there are questions over whether expenditure on offices, cars, staff expenses, and other trappings of status reduces shareholders' wealth or whether these are part of necessary performance incentives for executives. Some proponents of such expenditures believe that they serve to enhance contacts, breed confidence, improve the flow of information, and stimulate business. Third, if management asserts primacy of profits, this may in itself provide negative signals to employees about systems of corporate values. Where long-term success requires goodwill, commitment, and cooperation, focus on short-term profit may alienate or drive away those very employees upon whom long-term success depends.

Generally speaking, most companies turn over only about half of their earnings to stockholders as dividends. They plow the rest of their profits back into the operation. A major motivation of executives is to expand their operations faster than those of their competitors. The important point, however, is that without profit over the long term no firm can survive. For growing firms in competitive markets a major indicator of executive competence is the ability to augment company earnings by increasing sales or productivity or by achieving savings in other ways. This principle distinguishes the field of business from other fields. A drug company makes pharmaceuticals and may be interested in improving health, but it exists, first and foremost, to make profits. If it found that it could make more money by manufacturing frozen orange juice, it might choose to do so.

The modern executive. Much has been written about business executives as "organization men." According to this view, typical company managers no longer display the individualism of earlier generations of entrepreneurs. They seek protection in committee-made decisions and tailor their personalities to please their superiors; they aim to be good "team" members, adopting the firm's values as their own. The view is commonly held that there are companies—and entire industries—that have discouraged innovative ideas. The real question now is whether or not companies will develop policies to encourage autonomy and adventuresomeness among managers.

Company cadets in Japan

In Japan, where the employees of large corporations tend to remain with the same employer throughout their working lives, the corporations recruit young men upon their graduation from universities and train them as company cadets. Those among the cadets who demonstrate ability and a personality compatible with the organization are later selected as managers. Because of the seniority system, many are well past middle age before they achieve high status. There are signs that the system is weakening, however, as efforts are more often made to lift promising young men out of low-echelon positions. Criticism of the traditional method has been stimulated by the example of some of the newer corporations and of those owned by foreign capital. The few men in the Japanese business world who have emerged as personalities are either founders of corporations, managers of family enterprises, or small businessmen. They share a strong inclination to make their own decisions and to minimize the role of directors and boards. (S.N.W./Ed.)

Business finance

The planning, analysis, and control operations are responsibilities of the financial manager, who is usually close to the top of the organizational structure of a firm. Typically the manager is a member of the first level of the company staff in a large organization. In very large firms major financial decisions are often made by a finance committee. In small firms, the owner-manager usually conducts the financial operations. Much of the day-to-day work of business finance is conducted by lower level staff; their work includes handling cash receipts and disbursements, borrowing from commercial banks on a regular and continuing basis, and formulating cash budgets.

Financial decisions affect both the profitability and the riskiness of a firm's operations. An increase in the cash position, for instance, reduces risk; but, because cash is not an earning asset, converting other types of assets to cash reduces the firm's profitability. Similarly, the use of additional debt raises the profitability of the firm, but more debt means more risk. Striking a balance—between risk and profitability—that will maintain the long-term value of the firm's securities is the task of finance.

SHORT-TERM FINANCIAL OPERATIONS

Financial planning and control. Short-term financial operations are closely involved with the financial planning and control activities of the firm. These include financial ratio analysis, profit planning, financial forecasting, and budgeting.

Analysis of the balance sheet

Financial ratio analysis. A firm's balance sheet contains many items that taken by themselves have no clear meaning. Financial ratio analysis is a way of appraising their relative importance. The ratio of current assets to current liabilities, for example, gives the analyst an idea of the extent to which the firm can meet its current obligations. This is known as a liquidity ratio. Financial leverage ratios (such as the debt–asset ratio and debt as a percentage of total capitalization) are used to make judgments about the advantages to be gained from raising funds by the issuance of bonds (debt) rather than stock. Activity ratios, relating to the turnover of such asset categories as inventories, accounts receivable, and fixed assets, show how intensively the firm is employing its assets. The firm's primary operating objective is to earn a good return on its invested capital, and various profit ratios (profits as a percentage of sales, of assets, or of net worth) show how successfully it is meeting this objective.

Ratio analysis is used to compare a firm's performance with that of other firms in the same industry or with the performance of industry in general. It is also used to study trends in the firm's performance over time and thus to anticipate problems before they develop.

Profit planning. Ratio analysis applies to the firm's current operating posture. But the firm must also plan for future growth. This requires decisions as to the expansion of existing operations and, in manufacturing, to the development of new product lines. The firm must choose between productive processes requiring various degrees of mechanization or automation—that is, various amounts of fixed capital in the form of machinery and equipment. This will increase fixed costs—that is, costs that are relatively constant and do not decrease when the firm is operating at levels below full capacity. The higher the proportion of fixed costs to total costs, the higher must be the level of operation before profits begin, and the more sensitive will profits be to changes in the level of operation.

Estimate of sales

Financial forecasting. The financial manager must also make overall forecasts of future capital requirements to assure that funds will be available to finance new investment programs. The first step in making such a forecast is to obtain an estimate of sales during each year of the planning period. This estimate is worked out jointly by the marketing, production, and finance departments: the marketing manager estimates demand; the production manager estimates capacity; and the financial manager estimates availability of funds to finance new accounts receivable, inventories, and fixed assets.

For the predicted level of sales, the financial manager estimates the funds that will be available from the company's operations and compares this amount with what will be needed to pay for the new fixed assets (machinery, equipment, etc.). If the growth rate exceeds 10 percent a year, asset requirements are likely to exceed internal sources of funds, so plans must be made to finance them by issuing securities. If, on the other hand, growth is slow, more funds will be generated than are required to support the estimated growth in sales. In this case, the financial manager will consider a number of alternatives, including increasing dividends to stockholders, retiring debt, using excess funds to acquire other firms, or, perhaps, increasing expenditures on research and development.

Control through budgeting. Once the firm's general goals for the planning period have been established, the next step is to set up a detailed plan of operation—the budget. A complete budget system encompasses all aspects of the firm's operations over the planning period. It may even allow for changes in plans as required by factors outside the firm's control.

Budgeting is a part of the total planning activity in the firm, so it must begin with a statement of the firm's long-range plan. The long-range plan includes a long-range sales forecast. This forecast requires a determination of the number and types of products that will be manufactured in the years encompassed by the long-range plan. Short-term budgets are formulated within the framework of the long-range plan. Normally, there is a budget for every individual product and for every significant activity of the firm.

Establishing budgetary controls requires a realistic understanding of the firm's activities. For example, a small firm purchases more parts and uses more labour and less machinery; a larger firm will buy raw materials and use machinery to manufacture end items. In consequence, the smaller firm should budget higher parts and labour cost ratios, while the larger firm should budget higher overhead cost ratios and larger investments in fixed assets. If standards are unrealistically high, frustrations and resentment will develop. If standards are unduly lax, costs will be out of control, profits will suffer, and morale will drop.

The cash budget and management of cash. One of the principal methods of forecasting the financial needs of a business is the cash budget. The cash budget forecasts the combined effects of planned operations on the firm's cash flow. A positive net cash flow means that the firm will have surplus funds to invest. But if the cash budget indicates that an increase in the volume of operations will lead to a negative cash flow, additional financing will be required. The cash budget thus indicates the amount of funds that will be needed or available month by month, or even week by week.

Disposition of excess cash

A firm may have excess cash for a number of reasons. There are likely to be seasonal or cyclical fluctuations in business. Resources may be deliberately accumulated as a protection against a number of contingencies. Since it is wasteful to allow large amounts of cash to remain idle, the financial manager will try to find short-term investments for sums that he will need later. Short-term government or business securities can be selected and balanced in such a way that he obtains the maturities and risks appropriate to the financial situation of his firm.

Managing accounts receivable. Accounts receivable are the credit a firm gives its customers. The volume and terms of such credit vary among businesses and among nations; for manufacturing firms in the United States, for example, the ratio of receivables to sales ranges between 8 and 12 percent, representing an average collection period of approximately one month. The basis of a firm's credit policy is the practice in its industry; generally, a firm must meet the terms offered by competitors. Much depends, of course, on the individual customer's credit standing.

To evaluate a customer as a credit risk, the credit manager considers what may be called the five C's of credit: character, capacity, capital, collateral, and conditions. Information on these items is obtained from the firm's previous experience with the customer, supplemented by information from various credit associations and credit-reporting agencies. In reviewing a credit program, the financial manager should regard losses from bad debts as part of the cost of doing business. Accounts receivable represent an investment in the expansion of sales. The return of this investment can be calculated as in any capital budgeting problem.

Managing inventories. Every company must carry stocks of goods and materials in inventory. The size of the investment in inventory depends on various factors, including the level of sales, the nature of the productive processes, and the speed with which goods perish or become obsolescent.

The problems involved in managing inventories are basically the same as those in managing other assets, including cash. A basic stock must be on hand at all times. Because the unexpected may occur, it is also wise to have safety stocks; these represent the little extra needed to avoid the costs of not having enough. Additional amounts—anticipation stocks—may be required to meet future growth needs. Finally, some inventory accumulation results from the economies of purchasing in large quantities: it is always cheaper to buy more than is immediately needed, whether of raw materials, money, or plant and equipment. There is a standard procedure for determining the most economical amounts to order, one that relates purchasing requirements to costs and carrying charges. While carrying charges rise as average inventory holdings increase, certain other costs (ordering costs and stock-out costs) fall as average inventory holdings rise. These two sets of costs comprise the total cost of ordering and carrying inventories, and it is fairly easy to calculate an optimal order size that will minimize total inventory costs.

Working capital

Short-term financing. The main sources of short-term financing are (1) trade credit, (2) loans from commercial banks, and (3) commercial paper.

Trade credit. A firm customarily buys its supplies and materials on credit from other firms, recording the debt as an account payable. This trade credit, as it is commonly called, is the largest single category of short-term credit. Credit terms are usually expressed with a discount for prompt payment. Thus the seller may state that if payment is made within 10 days of the invoice date, a 2 percent cash discount will be allowed. If the cash discount is not taken, payment is due 30 days after the date of invoice. The cost of not taking cash discounts is the price of the credit, an amount equal to an effective rate of interest of 36.7 percent annually.

Commercial banks. Commercial bank lending appears on the balance sheet as notes payable and is second in importance to trade credit as a source of short-term financing. Banks occupy a pivotal position in the short-term and intermediate-term money markets. As a firm's financing needs grow, banks are called upon to provide additional funds. A single loan obtained from a bank by a business firm is not different in principle from a loan obtained by an individual. The firm signs a conventional promissory note. Repayment is made in a lump sum at maturity or in installments throughout the life of the loan. A line of credit, as distinguished from a single loan, is a formal or informal understanding between the bank and the borrower as to the maximum loan balance the bank will allow at any one time.

In the United States the role of commercial banks in the financing of business is somewhat unusual. The Banking Act of 1933 required the separation of investment banking and commercial banking functions, whereas in most other countries commercial banks are a source of long-term as well as of short-term financing.

Commercial paper. Commercial paper, a third source of short-term credit, consists of well-established firms' promissory notes sold primarily to other business firms, insurance companies, pension funds, and banks. Commercial paper is issued for periods varying from two to six months. The rates on prime commercial paper vary, but they are generally slightly below those on prime business loans.

A basic limitation of the commercial paper market is that its resources are limited to the excess liquidity that corporations, the main suppliers of funds, may have at any particular time. Another disadvantage is the impersonality of the dealings; a bank is much more likely to help a good customer weather a storm than is a commercial paper dealer.

Use of security in short-term financing. Most short-term business loans are unsecured. It is ordinarily better to borrow on an unsecured basis, but frequently a borrower's credit rating is not strong enough to justify an unsecured loan. The most common types of collateral used for short-term credit are inventories and accounts receivable.

Financing through accounts receivable can be done either by pledging the receivables or by selling them outright, a process called factoring in the United States. When a receivable is pledged, the borrower retains the risk that the person or firm who owes the receivable will not pay; this risk is typically passed on to the lender when factoring is involved.

When loans are secured by inventory, the lender takes title to them. He may or may not take physical possession of them. Under a field warehousing arrangement, the inventory is under the physical control of a warehouse company, which releases the inventory only on order from the lending institution. Canned goods, lumber, steel, coal, and other standardized products are the types of goods usually covered in field warehouse arrangements.

Financing equipment

Intermediate-term financing. Whereas short-term loans are repaid in a period of weeks or months, intermediate-term loans are scheduled for repayment in one to 15 years. Obligations due in 15 or more years are thought of as long-term debt. The major forms of intermediate-term financing include (1) term loans, (2) conditional sales contracts, and (3) lease financing.

Term loans. A term loan is a business credit with a maturity of more than one year but less than 15 years. Usually the term loan is retired by systematic repayments (amortization payments) over its life. It may be secured by a chattel mortgage on equipment, but the larger, stronger companies are able to borrow on an unsecured basis. Commercial banks and life insurance companies are the principal suppliers of term loans. The interest cost of term loans varies with the size of the loan and the strength of the borrower.

Term loans involve more risk to the lender than do short-term loans. The lender's funds are tied up for a long period, and during this time the borrower's situation can change markedly. To protect themselves, lenders often include in the loan agreement stipulations that the borrower will maintain his current liquidity ratio at a specified level, limit his acquisitions of fixed assets, keep his debt ratio below a stated amount, and in general follow policies that are acceptable to the lender.

Conditional sales contracts. Conditional sales contracts are a common method of obtaining equipment byagreeing to pay forit in installments over a period of up to five years. Until payment is completed, the seller of the equipment continues to hold title to the equipment.

Leasing. It is not necessary to purchase assets in order to use them. Railroad companies in the United States, for instance, have acquired much of their equipment by leasing it. Whether or not leasing is advantageous depends—aside from tax advantages—on the firm's access to funds; leasing provides an alternative method of financing. A lease contract, however, being a fixed obligation, is similar to debt and uses some of the firm's debt-carrying ability. It will generally be advantageous to a firm to own its land and buildings, because their value is likely to increase, but this consideration does not apply to equipment.

The statement is frequently made that leasing involves higher interest rates than other forms of financing, but this need not always be true. Much depends on the firm's standing as a credit risk. Moreover, it is difficult to separate the money costs of leasing from the other services that may be embodied in a leasing contract. If the leasing company can perform nonfinancial services, such as maintenance of the equipment, at a lower cost than the lessee or someone else could perform them, the effective cost of leasing may be lower than for funds obtained from borrowing or other sources.

Although leasing involves fixed charges, it enables a firm to present lower debt-to-asset ratios in its financial statements. Many lenders, in examining financial statements, give less weight to a lease obligation than to a loan obligation.

LONG-TERM FINANCIAL OPERATIONS

Managing the capital structure

Use of bonds. Long-term capital may be raised either through borrowing or by the issuance of stock. Long-term borrowing is done by selling bonds, which are promissory notes that obligate the firm to pay interest at specific times. Secured bondholders have prior claim on the firm's assets. If the company goes out of business, the bondholders are entitled to be paid the face value of their holdings plus interest. Stockholders, on the other hand, have no more than a residual claim on the company: they are entitled to a share of the profits, if there are any, but it is the prerogative of the board of directors to decide whether a dividend will be paid and how large it will be.

Long-term financing involves the choice between debt (bonds) and equity (stocks). Each firm will choose its own capital structure, seeking the combination of debt and equity that will minimize the costs of raising capital. As conditions in the capital market vary (for instance, changes in interest rates, the availability of funds, and the relative costs of alternative methods of financing), the firm's desired capital structure will change correspondingly.

The larger the proportion of debt in the capital structure (leverage), the higher will be the returns to equity. This is because bondholders do not share in the profits. The difficulty with this, of course, is that a high proportion of debt increases the firm's fixed costs and increases the degree of fluctuation in the returns to equity for any given degree of fluctuation in the level of sales. If used successfully, leverage increases the returns to owners but decreases the returns to owners when it is used unsuccessfully. Indeed, if leverage is unsuccessful, the result may be the bankruptcy of the firm.

Forms of long-term debt. There are various forms of long-term debt. A mortgage bond is one secured by a lien on fixed assets such as plant and equipment. A debenture is a bond not secured by specific assets but accepted by investors because the firm has a high credit standing or obligates itself to follow policies that assure a high rate of earnings. A still more junior lien is the subordinated debenture, which is secondary to all other debentures and specifically to short-term bank loans.

The use of long-term debt will be encouraged when sales and earnings are relatively stable, profit margins are large enough to make more leverage advantageous to the stockholders, an increase is expected in profits or the general price level, the existing debt ratio is relatively low, the price–earnings ratios on common stock are low in relation to the levels of interest rates, management is concerned with maintaining voting control over the company, cash-flow requirements under the bond agreement are not burdensome, and the restrictions on management written into the bond indenture are not onerous.

Stock. Equity financing is done with common and preferred stock. Preferred stock usually has priority over common stock with respect to earnings and claims on assets in liquidation. Preferred stock is usually cumulative—that is, the omission of dividends in one or more years creates an accumulated claim. The dividends on preferred stock are usually fixed at a specific percentage of face value. A company issuing preferred stock gains the advantages of limited dividends and no maturity—that is, the advantages of selling bonds but without the restrictions of bonds. Companies sell preferred stock when they seek more leverage but wish to avoid the fixed charges of debt. The advantages of preferred stock will be reinforced if a company's debt ratio is already high and if common stock financing is relatively expensive.

If a bond or preferred stock issue was sold when interest rates were higher than they are at present, it may be profitable to call the old issue and refund it with a new, lower cost issue. This depends on how the immediate costs and premiums that must be paid compare with the annual savings that can be obtained.

Earnings and dividend policies. The size and frequency of dividend payments are critical issues in company policy. Dividend policy affects the financial structure, the flow of funds, corporate liquidity, stock prices, and the morale of stockholders. If earnings are paid out as dividends, they cannot be used for company expansion. Some stockholders are interested in receiving maximum current returns on their investment, while others prefer reinvestment of earnings so that the company's capital will increase.

Companies tend to reinvest their earnings to a higher degree when there are chances for profitable expansion. Thus, at times when profits are high, the amounts reinvested are greater and dividends are smaller. For similar reasons, reinvestment is likely to decrease when profits decline, and dividends are likely to increase.

High-dividend policy

Companies having relatively stable earnings over a period of years tend to pay high dividends. Well-established large firms are likely to pay higher than average dividends because they have better access to capital markets and are not as dependent as other firms on internal financing. In the same way, if a firm has a strong cash or liquidity position, it is likely to pay higher dividends than if it has not. A firm with heavy indebtedness has implicitly committed itself to paying relatively low dividends and retaining earnings to service the debt. It may decide, however, to continue with high dividends in order to facilitate a stock issue or a refunding of its debt. If the directors of a company are concerned with maintaining control of it, they may retain earnings so that they can finance expansion without having to issue stock. Some companies favour a stable dividend policy rather than allow dividends to fluctuate with earnings; the dividend rate will then be lower when profits are high and higher when profits are temporarily in decline. Companies whose stock is closely held by a few high-income stockholders are likely to pay lower dividends in order to permit the stockholders to avoid an increase in personal income taxes.

In Europe, until recently, company financing tended to rely heavily on internal sources. This was because many companies were owned by families and, also, because a highly developed capital market was lacking. In the less developed countries today, firms rely heavily on internal financing; they also tend to make more use of short-term bank loans and other forms of short-term financing than is typical in other countries.

Convertible bonds and stock warrants. Companies sometimes issue bonds or preferred stock that give holders the option of converting them into common stock or of purchasing stock at favourable prices. Convertible bonds carry the option of conversion into common stock at a specified price during a particular period. Stock purchase warrants are given with bonds or preferred stock as an inducement to the investor; they permit him to buy common stock in the company at a stated price whenever he chooses. Such option privileges make it easier for small companies to sell bonds or preferred stock. They help large companies to float new issues on more favourable terms than they could otherwise obtain. When bondholders exercise conversion rights, the company's debt ratio is reduced because bonds are replaced by stock. The exercise of stock warrants, on the other hand, brings additional funds into the company but leaves the existing debt or preferred stock on the books. Option privileges also permit a company to sell new stock at more favourable prices than those prevailing at the time of issue, since the prices stated on the options are higher. They are most popular, therefore, at times when stock prices are expected to have an upward trend.

GROWTH AND DECLINE

Growth by combination

Mergers and consolidations. Companies often grow by combining with other companies. One company may purchase all or part of another; two companies may merge by exchanging shares; or a wholly new company may be formed through consolidation of the old companies. From the financial manager's viewpoint, this kind of expansion is like any other investment decision: the acquisition should be made if it increases the acquiring firm's net present value as reflected in the price of its stock.

The most important term that must be negotiated in a combination is the price the acquiring firm will pay for the assets it takes over. Present earnings, expected future earnings, and the effects of the merger on the rate of earnings growth of the surviving firm are perhaps the most important determinants of the price that will be paid. Current market prices are the second most important determinant of prices in mergers: depending on whether or not asset values are indicative of the earning power of the acquired firm, book values may exert an important influence on the terms of the merger. Other, nonmeasurable, factors are sometimes the overriding determinant in bringing companies together: synergistic effects (wherein the net result is greater than the combined value of the individual components) may be present to a sufficient extent to warrant paying more for the firm to be acquired than earnings and asset values would indicate.

A merger may be treated as either a purchase or a pooling of interests. In a purchase, a larger firm generally takes over a smaller one and assumes all management control. The amount actually paid for the smaller firm is reflected in the acquiring firm's balance sheet; if more was paid for the acquired firm than the book value of its assets, the difference is reflected in the acquiring firm's financial statements as goodwill. In a pooling of interests, the merged firms are usually about the same size; both managements carry on important functions after the merger; and common stock, rather than cash or bonds, is used in payment. The total assets of the surviving firm in a pooling are equal to the sum of the assets of the two independent companies, and no surplus remains to be written off as a charge against earnings.

Basis of a successful merger

The basic requirements for the success of a merger are that it fit into a soundly conceived long-range plan and that the resulting firm have performance characteristics superior to those attainable by the previous companies independently. In the heady environment of a rising stock market, mergers have often been motivated by superficial financial aims. Companies with stock selling at a high price relative to earnings have found it advantageous to merge with companies having a lower price–earnings ratio; this enables them to increase their earnings per share, thus appealing to investors who purchase stock on the basis of earnings.

Efficiencies in general management

Some mergers, particularly those of "conglomerates," which bring together firms in unrelated fields, owe their success to economies of management. In the 1950s certain large-scale changes in managerial technology occurred. The role of the general management functions (planning, control, organizing, information) and other functions centralized at top management levels (research, finance, legal services) increased in importance. The costs of managing large, diversified firms were substantially reduced. This explains generally the spread of conglomerate mergers in the United States.

In mergers, one firm disappears. An alternative is for one firm to buy all or a majority of the voting stock of another and to run it as an operating subsidiary. The acquiring firm is then called a holding company. There are several advantages in the holding company: it can control the acquired firm with a smaller investment than would be required in a merger; each firm remains a separate legal entity, and the obligations of one are separate from those of the other; and stockholder approval is not necessary, as it is in the case of a merger. There are also disadvantages to holding companies, including the possibility of multiple taxation and the danger that the high rate of leverage will amplify the earnings fluctuations of the operating companies.

Reorganization. When a firm cannot operate profitably the owners may seek to reorganize it. The first question to be answered is whether the firm might not be better off by ceasing to do business. If the decision is made that the firm is to survive, it must be put through the process of reorganization. Legal procedures are always costly, especially in the case of business failure; both the debtor and the creditors are better off if matters can be handled on an informal basis rather than through the courts. The informal procedures used in reorganization are (1) extension, which postpones the settlement of outstanding debt, and (2) composition, which reduces the amount owed.

If voluntary settlement through extension or composition is not possible, the matter must be taken to court. If the court decides on reorganization rather than liquidation, it appoints a trustee to control the firm and to prepare a formal plan of reorganization. The plan must meet standards of fairness and feasibility: the concept of fairness involves the appropriate distribution of proceeds to each claimant; the test of feasibility relates to the ability of the new enterprise to carry the fixed charges resulting from the reorganization plan. (J.F.W.)

Modern trends

Multinational and conglomerate companies

The sheer size of the largest limited-liability companies, or corporations—especially "multinationals," with holdings across the world—has been a subject of discussion and public concern since the end of the 19th century. For with this rise has come market and political power. While some large firms have declined, been taken over, or gone out of business, others have grown to replace them. The giant firms continue to increase their sales and assets, by expanding their markets, by diversifying, and by absorbing smaller companies. Diversification carried to the extreme has brought into being the conglomerate company, which acquires and operates subsidiaries that are often in unrelated fields. The holding company, with the conglomerate, acts as a kind of internal stock market, allocating funds to its subsidiaries on the basis of financial performance. The decline or failure of many conglomerates, however, has cast doubt upon the competence of any one group of executives to manage a diversity of unrelated operations. Empirical evidence from the United States suggests that conglomerates have been less successful financially than companies that have had a clear product-market focus based on organizational strengths and competencies.

The causes of such vast corporate growth have found varying explanations. One school of thought, most prominently represented by the U.S. economist John Kenneth Galbraith, sees growth as stemming from the imperatives of modern technology. Only a large firm can employ the range of talent needed for research and development in areas such as aerospace and nuclear energy. And only companies of this stature have the capacity for innovating industrial processes and entering international markets. Just as government has had to grow in order to meet new responsibilities, so have corporations found that producing for the contemporary economy calls for the intricate interaction of executives, experts, and extensive staffs of employees. While there is certainly room for small firms, the kinds of goods and services that the public seems to want increasingly require the resources that only a large company can master.

Others hold that the optimum size of the efficient firm is substantially smaller than many people believe. For instance George Romney, a former president of the U.S. company American Motors Corporation, contended that an automobile company could prosper and be profitable while producing only 200,000 cars a year. By this reasoning, most of the divisions of the huge U.S. General Motors Corporation could be established as separate companies. Some research has shown that profit rates in industries having a large number of smaller firms are just as high as in those in which a few big companies dominate a market. In this view, corporate expansion stems not from technological necessity but rather from an impulse to acquire or establish new subsidiaries or to branch out into new fields. The structures of most large corporations are really the equivalent of a congeries of semi-independent companies. In some cases these divisions compete against one another as if they were separately owned. The picture has been further complicated by growth across national boundaries, producing multinational companies, principally firms from western Europe and North America. Their enormous size and extent raise questions about their accountability and political and economic influence and power.

THE IMPACT OF THE LARGE COMPANY

Power and influence of giant corporations

While it is generally agreed that the power of large companies extends beyond the economic sphere, this influence is difficult to measure in any objective way. The processes of business entail at least some effort to ensure the sympathetic enactment and enforcement of legislation, since costs and earnings are affected by tax rates and government regulations. Companies and business groups send agents to local and national capitals and use such vehicles as advertising to enlist support for policies that they favour. Although, in many countries, companies may not legally contribute directly to candidates running for public office, their executives and stockholders may do so as individuals. Companies may, however, make payments to influence peddlers and contribute to committees working to pass or defeat legislative proposals. In practical terms, many lawmakers look upon companies as part of their constituency, although, if their districts depend on local plants, these lawmakers may be concerned more with preserving jobs than with protecting company profits. In any case, limited-liability companies are central institutions in society; it would be unrealistic to expect them to remain aloof from the political process that affects their operations, performance, and principles.

The decisions made by company managements have ramifications throughout society. In effect, companies can decide which parts of the country or even which parts of the world will prosper and which will decline by choosing where to locate their plants and other installations. The giant companies not only decide what to produce but also help to instill in their customers a desire for the amenities that the companies make available. To the extent that large firms provide employment, their personnel requirements determine the curricula of schools and universities. For these reasons, individuals' aspirations and dissatisfactions are likely to be influenced by large companies. This does not mean that large business firms can influence the public in any way they choose; it is simply that they are the only institutions available to perform certain functions. Automobiles, typewriters, frozen food, and electric toasters must come from company auspices if they are to be provided at all. Understanding this dependence as a given, companies tend to create an environment congenial to the conduct of their business.

THE SOCIAL ROLE OF THE LARGE COMPANY

Some company executives believe that their companies should act as "responsible" public institutions, holding power in trust for the community. Most companies engage in at least some public-service projects and make contributions to charities. A certain percentage of these donations can be deducted from a corporation's taxable income. Most of the donated money goes to private health, education, and welfare agencies, ranging from local hospital and charity funds to civil-rights groups and cultural institutions.

At the other extreme, it is generally agreed that companies should reject the notion that they have public duties, that society as a whole will be better off if companies maximize their profits, for this will expand employment, improve technology, raise living standards, and also provide individuals with more money to donate to causes of their own choosing. A cornerstone of this argument is that management has no right to withhold dividends. If stockholders wish to give gifts themselves, they should do so from their personal funds. On the other hand, some critics complain that large companies have been much too conservative in defining their responsibilities. Not only have most firms avoided public controversy, but also they have sought to reap public-relations benefits from every sum that they donate. Very few, say the critics, have made more than a token effort to promote minority hiring, provide day-care centres, or take on school dropouts and former convicts. Companies have also been charged with abandoning the central cities, profiting from military contracts, misrepresenting their merchandise, and investing in foreign countries governed by repressive regimes. A perennial indictment has been that profits, prices, and executive compensation are too high, while the wages and taxes paid by corporations are too low.

Consumer and environmental criticism

In the late 20th century a new school of critics emerged who stressed the social costs of the large company. They charged that automobiles, pharmaceuticals, and other products were badly designed and dangerous to their users. The new consumer movement, led by such figures as the U.S. lawyer Ralph Nader, was joined by environmental critics who pointed to the quantities of waste products released into streams and into the air. Local and national laws were passed in an effort to set higher standards of safety and to force companies to install antipollution devices. Not all of the critics understand that the costs of these measures are passed on to the consumer. If a nuclear power plant must have cooling towers so that it does not discharge heated water into an adjacent lake, for example, the extra equipment results in higher electricity bills. Most companies are hesitant to take such steps on their own initiative, fearing that they will need to raise prices without thereby increasing profits. Society, however, is already paying for the costs of traffic congestion, trash removal, and nutritional deficiencies. The prices charged by companies are far from reflecting the total impact that the manufacture and consumption of their products have upon human life.

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The History of the Byzantine Empire

The very name Byzantine illustrates the misconceptions to which the empire's history has often been subject, for its inhabitants would hardly have considered the term appropriate to themselves or to their state. Theirs was, in their view, none other than the Roman Empire, founded shortly before the beginning of the Christian Era by God's grace to unify his people in preparation for the coming of his Son. Proud of that Christian and Roman heritage, convinced that their earthly empire so nearly resembled the heavenly pattern that it could never change, they called themselves Romaioi, or Romans. Modern historians agree with them only in part. The term East Rome accurately described the political unit embracing the Eastern provinces of the old Roman Empire until 476, while there were yet two emperors. The same term may even be used until the last half of the 6th century, as long as men continued to act and think according to patterns not unlike those prevailing in an earlier Roman Empire. During these same centuries, nonetheless, there were changes so profound in their cumulative effect that after the 7th century state and society in the East differed markedly from their earlier forms. In an effort to recognize that distinction, historians traditionally have described the medieval empire as "Byzantine."

The latter term is derived from the name Byzantium, borne by a colony of ancient Greek foundation on the European side of the Bosporus, midway between the Mediterranean and the Black Sea; the city was, by virtue of its location, a natural transit point between Europe and Asia Minor (Anatolia). Refounded as the "new Rome" by the emperor Constantine in 330, it was endowed by him with the name Constantinople, the city of Constantine. The derivation from Byzantium is suggestive in that it emphasizes a central aspect of Byzantine civilization: the degree to which the empire's administrative and intellectual life found a focus at Constantinople from 330 to 1453, the year of the city's last and unsuccessful defense under the 11th (or 12th) Constantine. The circumstances of the last defense are suggestive, too, for in 1453 the ancient, medieval, and modern worlds seemed briefly to meet. The last Constantine fell in defense of the new Rome built by the first Constantine. Walls that had held firm in the early Middle Ages against German, Hun, Avar, Slav, and Arab were breached finally by modern artillery, in the mysteries of which European technicians had instructed the most successful of the Central Asian invaders: the Ottoman Turks.

The fortunes of the empire thus were intimately entwined with those of peoples whose achievements and failures constitute the medieval history of both Europe and Asia. Nor did hostility always characterize the relations between Byzantines and those whom they considered "barbarian." Even though the Byzantine intellectual firmly believed that civilization ended with the boundaries of his world, he opened it to the barbarian, provided that the latter (with his kin) would accept Baptism and render loyalty to the emperor. Thanks to the settlements that resulted from such policies, many a name, seemingly Greek, disguises another of different origin: Slavic, perhaps, or Turkish. Barbarian illiteracy, in consequence, obscures the early generations of more than one family destined to rise to prominence in the empire's military or civil service. Byzantium was a melting-pot society, characterized during its earlier centuries by a degree of social mobility that belies the stereotype, often applied to it, of an immobile, caste-ridden society.

A source of strength in the early Middle Ages, Byzantium's central geographical position served it ill after the 10th century. The conquests of that age presented new problems of organization and assimilation, and these the emperors had to confront at precisely the time when older questions of economic and social policy pressed for answers in a new and acute form. Satisfactory solutions were never found. Bitter ethnic and religious hostility marked the history of the empire's later centuries, weakening Byzantium in the face of new enemies descending upon it from east and west. The empire finally collapsed when its administrative structures could no longer support the burden of leadership thrust upon it by military conquests.

This article is divided into the following sections:

The empire to 867

THE ROMAN AND CHRISTIAN BACKGROUND

Unity and diversity in the late Roman Empire. The Roman Empire, the ancestor of the Byzantine, remarkably blended unity and diversity, the former being by far the better known since its constituents were the predominant features of Roman civilization. The common Latin language, the coinage, the "international" army of the Roman legions, the urban network, the law, and the Greco-

Roman heritage of civic culture loomed largest among those bonds that Augustus and his successors hoped would bring unity and peace to a Mediterranean world exhausted by centuries of civil war. To strengthen these sinews of imperial civilization, the emperors hoped that a lively and spontaneous trade might develop among the several provinces. At the pinnacle of this world stood the emperor himself, the man of wisdom who would shelter the state from whatever mishaps fortune had darkly hidden. The emperor alone could provide this protection since, as the embodiment of all the virtues, he possessed in perfection those qualities displayed only imperfectly by his individual subjects.

The Roman formula of combating fortune with reason and therewith assuring unity throughout the Mediterranean world worked surprisingly well in view of the pressures for disunity that time was to multiply. Conquest had brought regions of diverse background under Roman rule. The Eastern provinces were ancient and populous centres of that urban life that for millennia had defined the character of Mediterranean civilization. The Western provinces had only lately entered upon their own course of urban development under the not always tender ministrations of their Roman masters.

Each of the aspects of unity enumerated above had its other side. Not everyone understood or spoke Latin. Paralleling and sometimes influencing Roman law were local customs and practices, understandably tenacious by reason of their antiquity. Pagan temples, Jewish synagogues, and Christian baptisteries attest to the range of organized religions with which the official forms of the Roman state, including those of emperor worship, could not always peacefully coexist. And far from unifying the Roman world, economic growth often created self-sufficient units in the several regions, provinces, or great estates.

Greco-Roman civic tradition

Given the obstacles against which the masters of the Roman state struggled, it is altogether remarkable that Roman patriotism was ever more than an empty formula, that cultivated gentlemen from the Pillars of Hercules to the Black Sea were aware that they had "something" in common. This "something" might be defined as the Greco-Roman civic tradition in the widest sense of its institutional, intellectual, and emotional implications. Grateful for the conditions of peace that fostered it, men of wealth and culture dedicated their time and resources to glorifying that tradition through adornment of the cities that exemplified it and through education of the young who they hoped might perpetuate it.

Upon this world the barbarians descended after about AD 150. To protect the frontier against them, warrior emperors devoted whatever energies they could spare from the constant struggle to reassert control over provinces where local regimes emerged. In view of the ensuing warfare, the widespread incidence of disease, and the rapid turnover among the occupants of the imperial throne, it would be easy to assume that little was left of either the traditional fabric of Greco-Roman society or the bureaucratic structure designed to support it.

Neither assumption is accurate. Devastation was haphazard, and some regions suffered while others did not. In fact, the economy and society of the empire as a whole during that period was more diverse than it had ever been. Impelled by necessity or lured by profit, people moved from province to province. Social disorder opened avenues to eminence and wealth that the more stable order of an earlier age had closed to the talented and the ambitious. For personal and dynastic reasons, emperors favoured certain towns and provinces at the expense of others, and the erratic course of succession to the throne, coupled with a resulting constant change among the top administrative officials, largely deprived economic and social policies of recognizable consistency.

The reforms of Diocletian and Constantine. The definition of consistent policy in imperial affairs was the achievement of two great soldier-emperors, Diocletian (ruled 284–305) and Constantine I (sole emperor 324–337), who together ended a century of anarchy and refounded the Roman state. There are many similarities between them, not the least being the range of problems to which they addressed themselves: both had learned from the 3rd-century anarchy that one man alone and unaided could not hope to control the multiform Roman world and protect its frontiers; as soldiers, both considered reform of the army a prime necessity in an age that demanded the utmost mobility in striking power; both found the old Rome and Italy an unsatisfactory military base for the bulk of the imperial forces. Deeply influenced by the soldier's penchant for hierarchy, system, and order, a taste that they shared with many of their contemporaries as well as the emperors who preceded them, they were appalled by the lack of system and the disorder characteristic of the economy and the society in which they lived. Both, in consequence, were eager to refine and regularize certain desperate expedients that had been adopted by their rough military predecessors to conduct the affairs of the Roman state. Whatever their personal religious convictions, both, finally, believed that imperial affairs would not prosper unless the emperor's subjects worshiped the right gods in the right way.

Rule of succession

The means they adopted to achieve these ends differ so profoundly that one, Diocletian, looks to the past and ends the history of Rome; the other, Constantine, looks to the future and founds the history of Byzantium. Thus, in the matter of succession to the imperial office, Diocletian adopted precedents he could have found in the practices of the 2nd century AD. He associated with himself a coemperor, or Augustus. Each Augustus then adopted a young colleague, or Caesar, to share in the rule and eventually to succeed the senior partner. This rule of four, or tetrarchy, failed of its purpose, and Constantine replaced it with the dynastic principle of hereditary succession, a procedure generally followed in subsequent centuries. To divide administrative responsibilities, Constantine replaced the single praetorian prefect, who had traditionally exercised both military and civil functions in close proximity to the emperor, with regional prefects established in the provinces and enjoying civil authority alone. In the course of the 4th century, four great "regional prefectures" emerged from these Constantinian beginnings, and the practice of separating civil from military authority persisted until the 7th century.

Contrasts in other areas of imperial policy are equally striking. Diocletian persecuted Christians and sought to revive the ancestral religion. Constantine, a convert to the new faith, raised it to the status of a "permitted religion." Diocletian established his headquarters at Nicomedia, a city that never rose above the status of a provincial centre during the Middle Ages, while Constantinople, the city of Constantine's foundation, flourished mightily. Diocletian sought to bring order into the economy by controlling wages and prices and by initiating a currency reform based upon a new gold piece, the aureus, struck at the rate of 60 to the pound of gold. The controls failed and the aureus vanished, to be succeeded by Constantine's gold solidus. The latter piece, struck at the lighter weight of 72 to the gold pound, remained the standard for centuries. For whatever reason, in summary, Constantine's policies proved extraordinarily fruitful. Some of them—notably hereditary succession, the recognition of Christianity, the currency reform, and the foundation of the capital—determined in a lasting way the several aspects of Byzantine civilization with which they are associated.

Yet it would be a mistake to consider Constantine a revolutionary or to overlook those areas in which, rather than innovating, he followed precedent. Earlier emperors had sought to constrain groups of men to perform certain tasks that were deemed vital to the survival of the state but that proved unremunerative or repellent to those forced to assume the burden. Such tasks included the tillage of the soil, which was the work of the peasant, or *colonus;* the transport of cheap bulky goods to the metropolitan centres of Rome or Constantinople, which was the work of the shipmaster, or *navicularius;* and services rendered by the *curiales,* members of the municipal senate charged with the assessment and collection of local taxes. Constantine's laws in many instances extended or even rendered hereditary these enforced responsibilities, thus laying the foundations for the system of *collegia,* or hereditary state guilds, that was to be so noteworthy a feature of late-Roman social life. Of particular importance, he required the *colonus* (peasant) to remain in the locality to which the tax lists ascribed him.

THE 5TH CENTURY: PERSISTENCE OF GRECO-ROMAN CIVILIZATION IN THE EAST

Whether innovative or traditional, Constantine's measures determined the thrust and direction of imperial policy throughout the 4th century and into the 5th. The state of the empire in 395 may, in fact, be described in terms of the outcome of Constantine's work. The dynastic principle was established so firmly that the emperor who died in that year, Theodosius I, could bequeath the imperial office jointly to his sons, both of whom were young and incompetent: Arcadius in the East and Honorius in the West. Never again would one man rule over the full extent of the empire in both its halves. Constantinople had probably grown to a population of between 200,000 and 500,000; in the 5th century the emperors sought to restrain rather than promote its growth. After 391 Christianity was far more than one among many religions: from that year onward, imperial decree prohibited all forms of pagan cult, and the temples were closed. Imperial pressure was often manifest at the church councils of the 4th century, with the emperor assuming a role he was destined to fill again during the 5th century in defining and suppressing heresy.

Economic prosperity

Economic and social policies. The empire's economy had prospered in a spotty fashion. Certain provinces, or parts of provinces such as northern Italy, flourished commercially as well as agriculturally. Constantinople, in particular, influenced urban growth and the exploitation of agricultural frontiers. Balkan towns along the roads leading to the great city prospered, while others not so favoured languished and even disappeared. Untilled land in the hilly regions of northern Syria fell under the plow to supply foodstuffs for the masses of Constantinople. As the 4th century progressed, not only did Constantine's solidus remain indeed solid gold, but evidence drawn from a wide range of sources suggests that gold in any form was far more abundant than it had been for at least two centuries. It may be that new sources of supply for the precious metal had been discovered: these perhaps were in spoils plundered from pagan temples; or perhaps were from mines newly exploited in western Africa and newly available to the lands of the empire, thanks to the appearance of camel-driving nomads who transported the gold across the Sahara to the Mediterranean coastline of North Africa.

The extreme social mobility noted in the late 3rd and early 4th centuries seems less characteristic of the second half of the latter century. Certainly the emperors continued their efforts to bind men collectively to their socially necessary tasks, but the repetition of laws tying the *colonus* to his estate, the *navicularius* to his ship, and the *curialis* to his municipal senate suggests that these edicts had little effect. Indeed, it would be a mistake to conclude from such legislation that Roman society was universally and uniformly organized in castes determined in response to imperial orders. There was always a distinction between what an emperor wanted and what he could obtain, and, as the foregoing survey has suggested, there were distinctions among the provinces as well.

Even before the end of the first quarter of the 5th century, these provincial differences were visible; and, in no small degree, they help to explain the survival of imperial government and Greco-Roman civilization in the East while both eventually perished in the West. Throughout the Eastern provinces, population levels seem to have remained higher, and the emperors in Constantinople never had to search (at least until the 6th century) for men to fill the ranks of their armies. As might be expected in those eastern lands in which urban civilization was several centuries old, cities persisted and, with them, a merchant class and a monetary economy. Eastern merchants, known in the sources as Syrians, assumed the carrying trade between East and West, often establishing colonies in the beleaguered cities of the latter region.

Most important, the emperor in the East never lost access to, or control over, his sources of manpower and money. An older and probably more wealthy senatorial class, or aristocracy, in the West consolidated its great estates and assumed a form of protection or patronage over the labouring rural classes, depriving the state of desperately needed military and financial services. The senatorial class in the East seems to have been of more recent origin, its beginnings to be found among those favourites or parvenus who had followed Constantine to his new capital. By the early 5th century, their wealth seems to have been, individually, much less than the resources at the disposal of their Western counterparts; their estates were far more scattered and their rural dependents less numerous. They were thus less able to challenge the imperial will and less able to interpose themselves between the state, on the one hand, and its potential soldiers or taxpayers, on the other.

Battle of Adrianople

Relations with the barbarians. These differences between Eastern and Western social structures, together with certain geographical features, account for the different reception found by the Germanic invaders of the 4th and 5th centuries in East and West. Although the Germanic people had eddied about the Danube and Rhine frontiers of the empire since the 2nd century, their major inroads were made only in the latter half of the 4th century, when the ferocious Huns drove the Ostrogoths and Visigoths to seek refuge within the Danubian frontier of the empire. The initial interaction between Roman and barbarian was far from amicable; the Romans seemed to have exploited their unwelcome guests, and the Goths rose in anger, defeating an East Roman army at Adrianople in 378 and killing the Eastern emperor in command. Emperor Theodosius (ruled 384–395) adopted a different policy, granting the Goths lands and according them the legal status of allies, or *foederati,* who fought within the ranks of the Roman armies as autonomous units under their own leaders.

Neither in West nor East did Theodosius' policy of accommodation and alliance prove popular. The Goths, like most Germanic peoples with the exception of the Franks and the Lombards, had been converted to Arian Christianity, which the Catholic, or Orthodox, Romans considered a dangerous heresy. The warlike ways of the Germans found little favour with a senatorial aristocracy essentially pacifist in its outlook, and the early 5th century is marked in both halves of the empire by reactions against Germanic leaders in high office. At Constantinople in 400, for example, the citizens rose against the senior officer of the imperial guard (*magister militum*), Gainas, slaughtering him together with his Gothic followers. Although this particular revolt was, in many respects, less productive of immediate results than similar episodes in the West, and the Germanic leaders later reappeared in roles of command throughout the East, the latter acted thenceforth as individuals without the support of those nearly autonomous groups of soldiers that western barbarian commanders continued to enjoy.

Furthermore, the East made good use of its resources in gold, in native manpower, and in diplomacy, while quickly learning how best to play off one enemy against another. In the reign of Theodosius II (408–450), the Huns under their chieftain Attila received subsidies of gold that both kept them in a state of uneasy peace with the Eastern Empire and may have proved profitable to those merchants of Constantinople who traded with the barbarians. When Marcian (ruled 450–457) refused to continue the subsidies, Attila was diverted from revenge by the prospect of conquests in the West. He never returned to challenge the Eastern Empire, and, with his death in 453, his Hunnic empire fell apart. Both Marcian and his successor, Leo I (ruled 457–474), had ruled under the tutelage of Flavius Ardaburius, Aspar; but Leo resolved to challenge Aspar's pre-eminence and the influence of the Goths elsewhere in the empire by favouring the warlike Isaurians and their chieftain, Tarasicodissa, whom he married to the imperial princess, Ariadne. The Isaurian followers of Tarasicodissa, who was to survive a stormy reign as the Emperor Zeno (474–491), were rough mountain folk from southern Anatolia and culturally probably even more barbarous than the Goths or the other Germans. Yet, in that they were the subjects of the Roman emperor in the East, they were

Dissolution of the Western Empire

undoubtedly Romans and proved an effective instrument to counter the Gothic challenge at Constantinople. In the prefecture of Illyricum, Zeno ended the menace of Theodoric the Amal by persuading him (488) to venture with his Ostrogoths into Italy. The latter province lay in the hands of the German chieftain Odoacer, who in 476 had deposed Romulus Augustulus, the last Roman emperor in the West. Thus, by suggesting that Theodoric conquer Italy as his Ostrogothic kingdom, Zeno maintained at least a nominal supremacy in that western land while ridding the Eastern Empire of an unruly subordinate.

With Zeno's death and the accession of the Roman civil servant Anastasius I (ruled 491–518), Isaurian occupation of the imperial office ended, but it was not until 498 that the forces of the new emperor effectively took the measure of Isaurian resistance. After the victory of that year, the loyal subject of the Eastern Roman emperor could breathe easily: Isaurians had been used to beat Germans, but the wild mountain folk had, in their turn, failed to take permanent possession of the imperial office. Imperial authority had maintained its integrity in the East while the Western Empire had dissolved into a number of successor states: the Angles and Saxons had invaded Britain as early as 410; the Visigoths had possessed portions of Spain since 417, and the Vandals had entered Africa in 429; the Franks, under Clovis I, had begun their conquest of central and southern Gaul in 481; and Theodoric was destined to rule in Italy until 526.

The Monophysite heresy

Religious controversy. If ethnic hostility within the empire was less a menace around the year 500 than it had often been in the past, dissensions stemming from religious controversy seriously threatened imperial unity, and the political history of the next century cannot be understood without some examination of the so-called Monophysite heresy. It was the second great heresy in the Eastern Empire, the first having been the dispute occasioned by the teachings of the Alexandrian presbyter Arius, who, in an effort to maintain the uniqueness and majesty of God the Father, had taught that he alone had existed from eternity, while God the Son had been created in time. Thanks in part to imperial support, the Arian heresy had persisted throughout the 4th century and was definitively condemned only in 381 with promulgation of the doctrine that Father and Son were of one substance and thus coexistent.

If the Fathers of the 4th century quarreled over the relations between God the Father and God the Son, those of the 5th century faced the problem of defining the relationship of the two natures—the human and the divine—within God the Son, Christ Jesus. The theologians of Alexandria generally held that the divine and human natures were united indistinguishably, whereas those of Antioch taught that two natures coexisted separately in Christ, the latter being "the chosen vessel of the Godhead . . . the man born of Mary." In the course of the 5th century, these two contrasting theological positions became the subject of a struggle for supremacy among the rival sees of Constantinople, Alexandria, and Rome. Nestorius, patriarch of Constantinople in 428, adopted the Antiochene formula, which, in his hands, came to stress the human nature of Christ to the neglect of the divine. His opponents (first the Alexandrian patriarch, Cyril, and later Cyril's followers, Dioscorus and Eutyches) in reaction emphasized the single divine nature of Christ, the result of the Incarnation. Their belief in Monophysitism, or the one nature of Christ as God the Son, became extraordinarily popular throughout the provinces of Egypt and Syria. Rome, in the person of Pope Leo I, declared in contrast for Dyophysitism, a creed teaching that two natures, perfect and perfectly distinct, existed in the single person of Christ. At the Council of Chalcedon (451), the latter view triumphed thanks to the support of Constantinople, which changed its position and condemned both Nestorianism, or the emphasis on the human nature of Christ, and Monophysitism, or the belief in the single divine nature.

More important for the purposes of military and political history than the theological details of the conflict was the impact Monophysitism produced on the several regions of the Mediterranean world. Partly because it provided a formula to express resistance to Constantinople's imperial rule, Monophysitism persisted in Egypt and Syria. Until these two provinces were lost to Islām in the 7th century, each Eastern emperor had somehow to cope with their separatist tendencies as expressed in the heresy. He had either to take arms against Monophysitism and attempt to extirpate it by force, to formulate a creed that would somehow blend it with Dyophysitism, or frankly to adopt the heresy as his own belief. None of these three alternatives proved successful, and religious hostility was not the least of the disaffections that led Egypt and Syria to yield, rather readily, to the Arab conqueror. If ever the East Roman emperor was to reassert his authority in the West, he necessarily had to discover a formula that would satisfy Western orthodoxy while not alienating Eastern Monophysitism.

Anastasius I

The empire at the end of the 5th century. In the reign of Anastasius I (491–518), all these tendencies of the 5th century found their focus: the sense of Romanitas, which demanded a Roman rather than an Isaurian or a German emperor, the conflict between Orthodoxy and Monophysitism, and the persisting economic prosperity of the Eastern Roman Empire. Acclaimed and elected as the Roman and Orthodox emperor who would end both the hated hegemony of the Isaurians and the detested activity of the Monophysite heretics, Anastasius succeeded in the first of these objectives while failing in the second. While he defeated the Isaurians and transported many of them from their Anatolian homeland into Thrace, he gradually came to support the Monophysite heresy despite the professions of Orthodoxy he had made upon the occasion of his coronation. If his policies won him followers in Egypt and Syria, they alienated his Orthodox subjects and led, finally, to constant unrest and civil war.

Anastasius' economic policies were far more successful; if they did not provide the basis for the noteworthy achievements of the 6th century in military affairs and the gentler arts of civilization, they at least explain why the Eastern Empire prospered in those respects during the period in question. An inflation of the copper currency, prevailing since the age of Constantine, finally ended with welcome results for those members of the lower classes who conducted their operations in the base metal. Responsibility for the collection of municipal taxes was taken from the members of the local senate and assigned to agents of the praetorian prefect. Trade and industry were probably stimulated by the termination of the *chrysargyron,* a tax in gold paid by the urban classes. If, by way of compensating for the resulting loss to the state, the rural classes had then to pay the land tax in money rather than kind, the mere fact that gold could be presumed to be available in the countryside is a striking index of rural prosperity. In the East, the economic resurgence of the 4th century had persisted, and it is not surprising that Anastasius enriched the treasury to the extent of 320,000 pounds of gold during the course of his reign.

With such financial resources at their disposal, the Emperor's successors could reasonably hope to reassert Roman authority among the western Germanic successor states, provided they could accomplish two objectives: first, they must heal the religious discord among their subjects; second, they must protect the eastern frontier against the threat of Sāsānian Persia. Since there was, in fact, to be concurrent warfare on both fronts during the 6th century, some knowledge of the age-old rivalry between Rome and Persia is essential to an understanding of the problems confronted by the greatest among Anastasius' successors, Justinian I (ruled 527–565), as he undertook the conquest of the West.

The rise of Sāsānian Persia

In 224 the ancient Persian Empire had passed into the hands of a new dynasty, the Sāsānians, whose regime brought new life to the enfeebled state. Having assured firm control over the vast lands already subject to them, the Sāsānians took up anew the old struggle with Rome for northern Mesopotamia and its fortress cities of Edessa and Nisibis, lying between the Tigris and the Euphrates. In the course of the 4th century, new sources of hostility emerged as East Rome became a Christian empire. Partly by reaction, Sāsānian Persia strengthened the ecclesiastical organization that served its Zoroastrian religion; intolerance and persecution became the order of the day within Persia, and strife between the empires assumed something of the character of religious warfare. Hostilities were exacerbated when Armenia, lying to the north between the two realms, converted to Christianity and thus seemed to menace the religious integrity of Persia. If small-scale warfare during the 4th and 5th centuries rarely erupted into major expeditions, the threat to Rome nonetheless remained constant, demanding vigilance and the construction of satisfactory fortifications. By 518, the balance might be said to have tipped in the favour of Persia as it won away the cities of Theodosiopolis, Amida, and Nisibis.

THE 6TH CENTURY: FROM EAST ROME TO BYZANTIUM

The accession of Justinian I

The 6th century opened, in effect, with the death of Anastasius and the accession of the Balkan soldier who replaced him, Justin I (ruled 518–527). During most of Justin's reign, actual power lay in the hands of his nephew and successor, Justinian I. The following account of these more than 40 years of Justinian's effective rule is based upon the works of Justinian's contemporary, the historian Procopius. The latter wrote a laudatory account of the Emperor's military achievements in his *Polemon* (*Wars*) and coupled it in his *Anecdota* (*Secret History*) with a venomous threefold attack upon the Emperor's personal life, the character of the empress Theodora, and the conduct of the empire's internal administration. Justinian's reign may be divided into three periods: (1) an initial age of conquest and cultural achievement extending until the decade of the 540s; (2) 10 years of crisis and near disaster during the 540s; and (3) the last decade of the reign, in which mood, temper, and social realities more nearly resembled those to be found under Justinian's successors than those prevailing throughout the first years of his own reign.

After 550, it is possible to begin to speak of a medieval Byzantine, rather than an ancient East Roman, empire. Of the four traumas that eventually transformed the one into the other—namely, pestilence, warfare, social upheaval, and the Arab Muslim assault of the 630s—the first two were features of Justinian's reign.

The years of achievement to 540. Justinian is but one example of the civilizing magic that Constantinople often worked upon the heirs of those who ventured within its walls. Justin, the uncle, was a rude and illiterate soldier; Justinian, the nephew, was a cultivated gentleman, adept at theology, a mighty builder of churches, and a sponsor of the codification of Roman law. All these accomplishments are, in the deepest sense of the word, civilian, and it is easy to forget that Justinian's empire was almost constantly at war during his reign. The history of East Rome during that period illustrates, in classical fashion, how the impact of war can transform ideas and institutions alike.

The reign opened with external warfare and internal strife. From Lazica to the Arabian Desert, the Persian frontier blazed with action in a series of campaigns in which many of the generals later destined for fame in the West first demonstrated their capacities. The strength of the East Roman armies is revealed in the fact that, while containing Persian might, Justinian could nonetheless dispatch troops to attack the Huns in the Crimea and to maintain the Danubian frontier against a host of enemies. In 532 Justinian decided to abandon military operations in favour of diplomacy. He negotiated, at the cost of considerable tribute, an "Endless Peace" with the Persian king, Khosrow, which freed the Roman's hands for operations in another quarter of the globe. Thus Justinian succeeded in attaining the first of the objectives needed for reconquest in the West: peace in the East.

Even before his accession, Justinian had aided in the attainment of the second. Shortly after his proclamation as emperor, Justin had summoned a council of bishops at Constantinople. The council reversed the policies of Anastasius, accepted the orthodox formula of Chalcedon, and called for negotiations with the pope. Justinian had personally participated in the ensuing discussions, which restored communion between Rome and all the Eastern churches save Egypt. No longer could a barbarian king hope to maintain the loyalties of his Catholic subjects by persuading them that a Monophysite emperor ruled in the East.

The Nika revolt

In the same year of 532, Justinian survived a revolt in Constantinople, stemming from the Nika riot, which initially threatened his life no less than his throne but, in the event, only strengthened his position. To understand the course of events, it is essential to remember that Constantinople, like other great East Roman cities, had often to depend upon its urban militia, or demes, to defend its walls. Coinciding with divisions within the demes were factions organized to support rival charioteers competing in the horse races: the Blues and the Greens. It was originally thought that these two factions were divided by differing political and religious views and that these views were aired to the Emperor during the races. More recent scholarship has shown that the factions were seldom motivated by anything higher than partisan fanaticism for their respective charioteers. The Nika riot—"*Nika!*" ("Conquer!" or "Win!") being the slogan shouted during the races—of 532, however, was one of the rare occasions when the factions voiced political opposition to the imperial government. Angered at the severity with which the urban prefect had suppressed a riot, the Blues and Greens first united and freed their leaders from prison; they insisted then that Justinian dismiss from the office two of his most unpopular officials: John of Cappadocia and Tribonian. Even though the Emperor yielded to their demands, the crowd was not appeased, converted its riot into a revolt, and proclaimed a nephew of Anastasius as emperor. Justinian was saved only because the empress, Theodora, refused to yield. Justinian's able general, Belisarius, sequestered the rebels in the Hippodrome and slaughtered them to the number of 30,000. The leaders were executed, and their estates passed, at least temporarily, into the Emperor's hands.

After 532 Justinian ruled more firmly than ever before. With the subsequent proclamation of the "Endless Peace," he could hope to use his earlier won reputation as a champion of Chalcedonian orthodoxy and appeal to those Western Romans who preferred the rule of a Catholic Roman emperor to that of an Arian German kinglet. In these early years of the 530s, Justinian could indeed pose as the pattern of a Roman and Christian emperor. Latin was his language, and his knowledge of Roman history and antiquities was profound.

Code of Justinian

In 529 his officials had completed a major collection of the emperors' laws and decrees promulgated since the reign of Hadrian. Called the Codex Constitutionum and partly founded upon the 5th-century Theodosian Code, it comprised the first of four works compiled between 529 and 565 called the Corpus Juris Civilis (Body of Civil Law), commonly known as the Code of Justinian. This first collection of imperial edicts, however, pales before the Digesta completed under Tribonian's direction in 533. In the latter work, order and system were found in (or forced upon) the contradictory rulings of the great Roman jurists; to facilitate instruction in the schools of law, a textbook, the Institutiones (533), was designed to accompany the Digesta. The fourth book, the Novellae Constitutiones Post Codicem (commonly called the Novels), consists of collections of Justinian's edicts promulgated between 534 and 565.

Meanwhile, architects and builders worked apace to complete the new Church of the Holy Wisdom, Hagia Sophia, designed to replace an older church destroyed in the course of the Nika revolt. In five years they had constructed the edifice, and it stands today as one of the major monuments of architectural history.

In 533 the moment had clearly come to reassert Christian Roman authority in the West, and Vandal North Africa seemed the most promising theatre of operations. Although a major expedition mounted under Leo I had failed to win back the province, political conditions in the Vandal monarchy had altered to the Eastern emperor's favour. When King Hilderich was deposed and replaced, Justinian could rightfully protest this action taken against a monarch who had ceased persecution of North African Catholics and had allied himself with Constantinople. The Eastern merchants favoured military action in the West, but Justinian's generals were reluctant; possibly for that reason, only a small force was dispatched under Belisarius. Success came with surprising ease after two engagements, and in 534 Justinian set about organizing this new addition to the provinces of the Roman Empire.

These were, in fact, years of major provincial reorganization, and not in North Africa alone. A series of edicts dated in 535 and 536, clearly conceived as part of a master plan by the prefect John of Cappadocia, altered administrative, judicial, and military structures in Thrace and Asia Minor. In general, John sought to provide a simplified and economical administrative structure in which overlapping jurisdictions were abolished, civil and military functions were sometimes combined in violation of Constantinian principles, and a reduced number of officials were provided with greater salaries to secure better personnel and to end the lure of bribery.

In the prefaces to his edicts, Justinian boasted of his reconstituted authority in North Africa, hinted at greater conquests to come, and—in return for the benefits his decrees were to provide—urged his subjects to pay their taxes promptly so that there might be "one harmony between ruler and ruled." Quite clearly the Emperor was organizing the state for the most strenuous military effort, and, later (possibly in 539), reforms were extended to Egypt, whence the export of grain was absolutely essential for the support of expeditionary armies and Constantinople.

Campaigns in Italy

Developments during 534 and 535 in Ostrogothic Italy made it the most likely victim after the fall of Vandal North Africa. When Theodoric died in 526, he was succeeded by a minor grandson for whom Theodoric's daughter, Amalasuntha, acted as regent. Upon the boy's death, Amalasuntha attempted to seize power in her own right and connived at the assassination of three of her chief enemies. Her diplomatic relations with the Eastern emperor had always been marked by cordiality and even dependency; thus, when Amalasuntha, in turn, met her death in a blood feud mounted by the slain men's families, Justinian seized the opportunity to protest the murder.

In 535, as in 533, a small, tentative expedition sent to the West—in this instance, to Sicily—met with easy success. At first the Goths negotiated; then they stiffened their resistance, deposed their king, Theodahad, in favour of a stronger man, Witigis, and attempted to block Belisarius' armies as they entered the Italian peninsula. There the progress of East Roman arms proved slower, and victory did not come until 540 when Belisarius captured Ravenna, the last major stronghold in the north, and, with it, King Witigis, a number of Gothic nobles, and the royal treasure.

All were dispatched to Constantinople, where Justinian was presumably thankful for the termination of hostilities in the West. Throughout the 530s, Justinian's generals almost constantly had to fight to preserve imperial authority in the new province of North Africa and in the Balkans as well. In 539 a Gothic embassy reached Persia, and the information it provided caused the king, Khosrow, to grow restive under the constraints of the "Endless Peace." During the next year (the same year [540] that a Bulgar force raided Macedonia and reached the long walls of Constantinople), Khosrow's armies reached even Antioch in the pursuit of booty and blackmail. They returned unhurt, and 541 witnessed the Persian capture of a fortress in Lazica. In Italy, meanwhile, the Goths chose a new king, Totila, under whose able leadership the military situation in that land was soon to be transformed.

Outbreak of bubonic plague

The crisis of midcentury. At last the menace of simultaneous war on two fronts threatened Justinian's plans. During the 550s, his armies were to prove equal to the challenge, but a major disaster prevented them from so doing between 541 and about 548. The disaster was the bubonic plague of 541–543, the first of those shocks, or traumas, mentioned earlier, that would eventually transform East Rome into the medieval Byzantine Empire. The plague was first noted in Egypt, and from there it passed through Syria and Asia Minor to Constantinople. By 543 it had reached Italy and Africa, and it may also have attacked the Persian armies on campaign in that year. In East Asia the disease has persisted into the 20th century, providing medical science with an opportunity to view its causes and course. Transmitted to humans by fleas from infected rodents, the plague attacks the glands and early manifests itself by swellings (buboes) in armpit and groin, whence the name bubonic. To judge from Procopius' description of its symptoms at Constantinople in 542, the disease then appeared in its more virulent pneumonic form, wherein the bacilli settle in the lungs of the victims. The appearance of the pneumonic form was particularly ominous because it may be transmitted directly from person to person, spreading the infection all the more readily and producing exceptionally high mortality rates. Comparative studies, based upon statistics derived from incidence of the same disease in late-medieval Europe, suggest that between one-third and one-half the population of Constantinople may well have died, while the lesser cities of the empire and the countryside by no means remained immune.

The short-term impact of the plague may be seen in several forms of human activity during the 540s. Justinian's legislation of those years is understandably preoccupied with wills and intestate succession. Labour was scarce, and workers demanded wages so high that Justinian sought to control them by edict, as the monarchs of France and England were to do during the plague of the 14th century. In military affairs, above all, the record of those years is one of defeat, stagnation, and missed opportunities. Rather than effective Roman opposition, it was Khosrow's own weariness of an unprofitable war that led him to sign a treaty of peace in 545, accepting tribute from Justinian and preserving Persian conquests in Lazica. Huns, Sclaveni, Antae, and Bulgars ravaged Thrace and Illyricum, meeting only slight opposition from Roman armies. In Africa a garrison diminished by plague nervously faced the threat of Moorish invasion. In Italy, Totila took the offensive, capturing southern Italy and Naples and even forcing his way into Rome (546) despite Belisarius' efforts to relieve the siege. Desperately, Justinian's great general called for reinforcements from the East; if ever they came, they were slow in arriving and proved numerically less than adequate to the task confronting them.

The last years of Justinian I. After about 548, Roman fortunes improved, and, by the mid-550s, Justinian had won victories in most theatres of operation, with the notable and ominous exception of the Balkans. A tour of the frontiers might begin with the East. In 551 the fortress of Petra was recovered from the Persians, but fighting continued in Lazica until a 50 years' peace, signed in 561, defined relations between the two great empires. On balance, the advantage lay with Justinian. Although Justinian agreed to continue payment of tribute in the amount of 30,000 solidi a year, Khosrow, in return, abandoned his claims to Lazica and undertook not to persecute his Christian subjects.

The silk trade

The treaty also regulated trade between Rome and Persia, since rivalry between the two great powers had always had its economic aspects, focused primarily upon the silk trade. Raw silk reached Constantinople through Persian intermediaries, either by a land route leading from China through Persia or by the agency of Persian merchants in the Indian Ocean. The need to break this Persian monopoly had led Justinian to search for new routes and new peoples to serve as intermediaries: in the south, the Ethiopian merchants of the kingdom of Aksum; in the north, the peoples around the Crimea and in the Caucasian kingdom of Lazica, as well as the Turks of the steppes beyond the Black Sea. Other valuable commodities were exchanged in the Black Sea region, including textiles, jewelry, and wine from East Rome for the furs, leather, and slaves offered by the barbarians; yet, silk remained the commodity of prime interest. It was fortunate, then, that before 561, East Roman agents had smuggled silkworms from China into Constantinople, establishing a silk industry that would liberate the empire from dependence on Persia and become one of medieval Byzantium's most important economic operations.

In the West, Justinian's successes were even more spectacular. By 550 the Moorish threat had ended in North Africa. In 552 the armies of Justinian had intervened in a quarrel among the Visigothic rulers of Spain, and the

East Roman troops overstayed the invitation extended them, seizing the opportunity to occupy on a more permanent basis certain towns in the southeastern corner of the Iberian Peninsula. Most important of all, Italy was recovered. Early in the 550s, Justinian assembled a vast army composed not only of Romans but also of barbarians, including Lombards, Heruli, and Gepids, as well as Persian deserters. Command of this host eventually was given to an unlikely but, as events were to prove, able commander: the eunuch and chamberlain Narses. In two decisive battles (Busta Gallorum and Mons Lactarius), the East Roman general defeated first Totila and then his successor, Teias. The Goths agreed to leave Italy. Despite the continued resistance of certain Gothic garrisons, coupled with the intervention of Franks and Alamanni, after 554 the land was essentially a province of the East Roman Empire.

Balkan affairs

In view of the wide mixture of peoples that descended upon it, the Balkans present a far more complex situation, and the Romans used a wider variety of tactics to contain the barbarians. After the Kutrigur Bulgar attack of 540, Justinian worked to extend a system of fortifications that ran in three zones through the Balkans and as far south as the Pass of Thermopylae. Fortresses, strongholds, and watchtowers were not, however, enough. The Slavs plundered Thrace in 545 and returned in 548 to menace Dyrrhachium; in 550 the Sclaveni, a Slavic people, reached a point about 40 miles (65 kilometres) from Constantinople. The major invasion came in 559, when the Kutrigur Bulgars, accompanied by Sclaveni, crossed the Danube and divided their force into three columns. One column reached Thermopylae; the second gained a foothold on the Gallipoli Peninsula near Constantinople; and the third advanced as far as the suburbs of Constantinople itself, which the aged Belisarius had to defend with an unlikely force of civilians, demesmen, and a few veterans. Worried by Roman naval action on the Danube, which seemed to menace the escape route home, the Kutrigurs broke off the attack, returned north, and found themselves under attack from the Utigurs, a people whose support Justinian's agents had earlier connived at and won by suitable bribes. The two peoples weakened each other in warfare, of which the episode of 559 was not the first instance, and this was precisely the result at which Byzantine diplomacy was aimed.

As long as the financial resources remained adequate, diplomacy proved the most satisfactory weapon in an age when military manpower was a scarce and precious commodity. Justinian's subordinates were to perfect it in their relationships with Balkan and south Russian peoples. For, if the Central Asian lands constituted a great reservoir of people, whence a new menace constantly emerged, the very proliferation of enemies meant that one might be used against another through skillful combination of bribery, treaty, and perfidy. East Roman relations in the late 6th century with the Avars, a Mongol people seeking refuge from the Turks, provide an excellent example of this "defensive imperialism." The Avar ambassadors reached Constantinople in 557, and, although they did not receive the lands they demanded, they were loaded with precious gifts and allied by treaty with the empire. The Avars moved westward from south Russia, subjugating Utigurs, Kutrigurs, and Slavic peoples to the profit of the empire. At the end of Justinian's reign, they stood on the Danube, a nomadic people hungry for lands and additional subsidies and by no means unskilled themselves in a sort of perfidious diplomacy that would help them pursue their objectives.

Impact of the bubonic plague

No summary of the quiet, but ominous, last years of Justinian's reign would be complete without some notice of the continuing attacks of bubonic plague and the impact they were to continue to produce until the 8th century. As have other societies subjected to devastation from warfare or disease, East Roman society might have compensated for its losses of the 540s had the survivors married early and produced more children in the succeeding generations. Two developments prevented recovery. Monasticism, with its demands for celibacy, grew apace in the 6th century, and the plague returned sporadically to attack those infants who might have replaced fallen members of the older generations.

The resulting shortage of manpower affected several aspects of a state and society that perceptibly were losing their Roman character and assuming their Byzantine. The construction of new churches, so noteworthy a feature of the earlier years, ceased as men did little more than rebuild or add to existing structures. An increasing need for taxes, together with a decreasing number of taxpayers, evoked stringent laws that forced members of a village tax group to assume collective responsibility for vacant or unproductive lands. This, contemporary sources avow, was a burden difficult to assume, in view of the shortage

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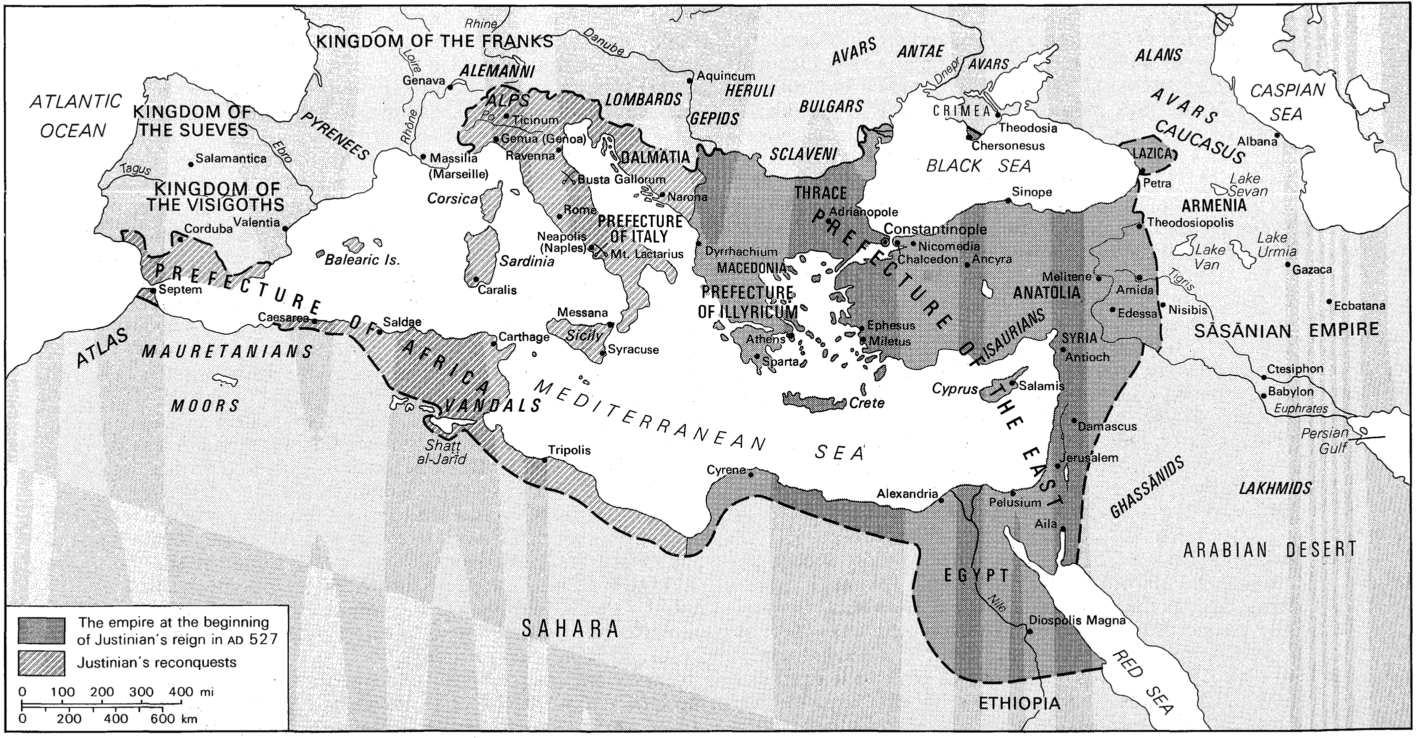

The Byzantine Empire at the death of Justinian in AD 565.

of agricultural workers after the plague. Finally, the armies that won the victories described above in east and west were largely victorious only because Justinian manned them as never before with barbarians: Goths, Armenians, Heruli, Gepids, Saracens, and Persians—to name only the most prominent. It was far from easy to maintain discipline among so motley an army; yet, once the unruly barbarian accepted the quieter life of the garrison soldier, he tended to lose his fighting capacity and prove, once the test came, of little value against the still warlike barbarian facing him beyond the frontier. The army, in short, was a creation of war and kept its quality only by participating in battlefield action, but further expansive warfare could hardly be undertaken by a society chronically short of men and money.

Problems of the late 6th century

In summary, the East Roman (or better, the Byzantine) state of the late 6th century seemed to confront many of the same threats that had destroyed the Western Empire in the 5th century. Barbarians pressed upon it from beyond the Balkan frontier, and peoples of barbarian origin manned the armies defending it. Wealth accumulated during the 5th century had been expended; and, to satisfy the basic economic and military needs of state and society, there were too few native Romans. If the Byzantine Empire avoided the fate of West Rome, it did so only because it was to combine valour and good luck with certain advantages of institutions, emotions, and attitudes that the older empire had failed to enjoy. One advantage already described, diplomatic skill, blends institutional and attitudinal change, for diplomacy would never have succeeded had not the Byzantine statesmen been far more curious and knowing than Justinian's 5th-century predecessor about the habits, customs, and movements of the barbarian peoples. The Byzantine's attitude had changed in yet another way. He was willing to accept the barbarian within his society provided that the latter, in his turn, accept orthodox Christianity and the emperor's authority. Christianity was often, to be sure, a veneer that cracked in moments of crisis, permitting a very old paganism to emerge, while loyalty to the emperor could be forsworn and often was. Despite these shortcomings, the Christian faith and the ecclesiastical institutions defined in the 6th century proved better instruments by far to unite men and stimulate their morale than the pagan literary culture of the Greco-Roman world.

Christian culture of the Byzantine Empire. Justinian's legislation dealt with almost every aspect of the Christian life: entrance into it by conversion and Baptism; administration of the sacraments that marked its several stages; proper conduct of the laity to avoid the wrath God would surely visit upon a sinful people; finally, the standards to be followed by those who lived the particularly holy life of the secular or monastic clergy. Pagans were ordered to attend church and accept Baptism, while a purge thinned their ranks in Constantinople, and masses of them were converted by missionaries in Asia Minor. Only the orthodox wife might enjoy the privileges of her dowry; Jews and Samaritans were denied, in addition to other civil disabilities, the privilege of testamentary inheritance unless they converted. A woman who worked as an actress might better serve God were she to forswear any oath she had taken, even though before God, to remain in that immoral profession. Blasphemy and sacrilege were forbidden, lest famine, earthquake, and pestilence punish the Christian society. Surely God would take vengeance upon Constantinople, as he had upon Sodom and Gomorrah, should the homosexual persist in his "unnatural" ways.

Byzantine morality

Justinian regulated the size of churches and monasteries, forbade them to profit from the sale of property, and complained of those priests and bishops who were unlearned in the forms of the liturgy. His efforts to improve the quality of the secular clergy, or those who conducted the affairs of the church in the world, were most opportune. The best possible men were needed, for, in most East Roman cities during the 6th century, imperial and civic officials gradually resigned many of their functions to the bishop, or patriarch. The latter collected taxes, dispensed justice, provided charity, organized commerce, negotiated with barbarians, and even mustered the soldiers. By the early 7th century, the typical Byzantine city, viewed from without, actually or potentially resembled a fortress; viewed from within, it was essentially a religious community under ecclesiastical leadership. Nor did Justinian neglect the monastic clergy, or those who had removed themselves from the world. Drawing upon the regulations to be found in the writings of the 4th-century Church Father St. Basil of Caesarea, as well as the acts of 4th- and 5th-century church councils, he ordered the cenobitic (or collective) form of monastic life in a fashion so minute that later codes, including the rule of St. Theodore the Studite in the 9th century, only develop the Justinianic foundation.

Probably the least successful of Justinian's ecclesiastical policies were those adopted in an attempt to reconcile Monophysites and orthodox Chalcedonians. After the success of negotiations that had done so much to conciliate the West during the reign of Justin I, Justinian attempted to win over the moderate Monophysites, separating them from the extremists. Of the complicated series of events that ensued, only the results need be noted. In developing a creed acceptable to the moderate Monophysites of the East, Justinian alienated the Chalcedonians of the West and thus sacrificed his earlier gains in that quarter. The extreme Monophysites refused to yield. Reacting against Justinian's persecutions, they strengthened their own ecclesiastical organization, with the result that many of the fortress cities noted above, especially those of Egypt and Syria, owed allegiance to Monophysite ecclesiastical leadership. To his successors, then, Justinian bequeathed the same religious problem he had inherited from Anastasius.

If, in contrast, his regulation of the Christian life proved successful, it was largely because his subjects themselves were ready to accept it. Traditional Greco-Roman culture was, to be sure, surprisingly tenacious and even productive during the 6th century and was always to remain the treasured possession of an intellectual elite in Byzantium; but the same century witnessed the growth of a Christian culture to rival it. Magnificent hymns written by St. Romanos Melodos mark the striking development of the liturgy during Justinian's reign, a development that was not without its social implications. Whereas traditional pagan culture was literary and its pursuit or enjoyment thereby limited to the leisured and wealthy, the Christian liturgical celebration and its musical component were available to all, regardless of place or position. Biography, too, became both markedly Christian and markedly popular. Throughout the countryside and the city, holy men appeared in legend or in fact, exorcising demons, healing the sick, feeding the hungry, and warding off the invader. Following the pattern used in the 4th century by Athanasius to write the life of St. Anthony, hagiographers recorded the deeds of these extraordinary men, creating in the saint's life a form of literature that began to flower in the 6th and 7th centuries.

The rise of popular Christian culture

The vitality and pervasiveness of popular Christian culture manifested themselves most strongly in the veneration increasingly accorded the icon, an abstract and simplified image of Christ, the Virgin, or the saints. Notable for the timeless quality that its setting suggested and for the power expressed in the eyes of its subject, the icon seemingly violated the Second Commandment's explicit injunction against the veneration of any religious images. Since many in the early centuries of the church so believed, and in the 8th century the image breakers, or iconoclasts, were to adopt similar views, hostility toward images was nearly as tenacious an aspect of Christianity as it had been of Judaism before it.

The contrasting view—a willingness to accept images as a normal feature of Christian practice—would not have prevailed had it not satisfied certain powerful needs as Christianity spread among Gentiles long accustomed to representations of the divinity and among Hellenized Jews who had themselves earlier broken with the Mosaic commandment. The convert all the more readily accepted use of the image if he had brought into his Christianity, as many did, a heritage of Neoplatonism. The latter school taught that, through contemplation of that which could be seen (*i.e.*, the image of Christ), the mind might rise to contemplation of that which could not be seen (*i.e.,* the essence of Christ). From a belief that the seen suggests the unseen, it is but a short step to a belief that the seen contains the unseen and that the image deserves veneration because divine power somehow resides in it.

Men of the 4th century were encouraged to take such a step, influenced as they were by the analogous veneration that the Romans had long accorded the image of the emperor. Although the first Christians rejected this practice of their pagan contemporaries and refused to adore the image of a pagan emperor, their successors of the 4th century were less hesitant to render such honour to the images of the Christian emperors following Constantine. Since the emperor was God's vicegerent on Earth and his empire reflected the heavenly realm, the Christian must venerate, to an equal or greater degree, Christ and his saints. Thus the Second Commandment finally lost much of its force. Icons appeared in both private and public use during the last half of the 6th century: as a channel of divinity for the individual and as a talisman to guarantee success in battle. During the dark years following the end of Justinian I's reign, no other element of popular Christian belief better stimulated that high morale without which the Byzantine Empire would not have survived.

Justin II

The successors of Justinian: 565–610. Until Heraclius arrived to save the empire in 610, inconsistency and contradiction marked the policies adopted by the emperors, a reflection of their inability to solve the problems Justinian had bequeathed his successors. Justin II (565–578) haughtily refused to continue the payment of tribute to Avar or Persian; he thereby preserved the resources of the treasury, which he further increased by levying new taxes. Praiseworthy as his refusal to submit to blackmail may seem, Justin's intransigence only increased the menace to the empire. His successor, Tiberius II (578–582), removed the taxes and, choosing between his enemies, awarded subsidies to the Avars while taking military action against the Persians. Although Tiberius' general, Maurice, led an effective campaign on the eastern frontier, subsidies failed to restrain the Avars. They captured the Balkan fortress of Sirmium in 582, while the Slavs began inroads across the Danube that would take them, within 50 years, into Macedonia, Thrace, and Greece.

The accession of Maurice in 582 inaugurated a reign of 20 years marked by success against Persia, a reorganization of Byzantine government in the West, and the practice of economies during his Balkan campaigns that, however unavoidable, would destroy him in 602. Byzantine efforts against Sāsānian Persia were rewarded in 591 by a fortunate accident. The lawful claimant to the Persian throne, Khosrow II, appealed to Maurice for aid against the rebels who had challenged his succession. In gratitude for this support, Khosrow abandoned the frontier cities and the claims to Armenia, the two major sources of contention between Byzantium and Persia. The terms of the treaty gave Byzantium access, in Armenia, to a land rich in the soldiers it desperately needed and, equally important, an opportunity to concentrate on other frontiers where the situation had worsened.

The exarchate of Carthage and Ravenna

Confronted by a Visigothic resurgence in Spain and by the results of a Lombard invasion of Italy (568), which was steadily confining Byzantine power to Ravenna, Venice, and Calabria-Sicily in the south, Maurice developed a form of military government throughout the relatively secure province of North Africa and in whatever regions were left in Italy. He abandoned the old principle of separating civil from military powers, placing both in the hands of the generals, or exarchs, located, respectively, at Carthage and Ravenna. Their provinces, or exarchates, were subdivided into duchies composed of garrison centres that were manned not by professional soldiers but by conscript local landholders. The exarchate system of military government seems to have worked well: North Africa was generally quiet despite Moorish threats; and in 597 the ailing Maurice had intended to install his second son as emperor throughout those western possessions in which he had clearly not lost interest.

But the major thrust of his efforts during the last years of his reign was to be found in the Balkans, where, by dint of constant campaigning, his armies had forced the Avars back across the Danube by 602. In the course of these military operations, Maurice made two mistakes: the first weakened him; the second destroyed him together with his dynasty. Rather than constantly accompanying his armies in the field, as his 7th- and 8th-century successors were to do, Maurice remained for the most part in Constantinople, losing an opportunity to engage the personal loyalty of his troops. He could not count on their obedience when he issued unwelcome commands from afar that decreased their pay in 588, ordered them to accept uniforms and weapons in kind rather than in cash equivalents, and, in 602, required the soldiers to establish winter quarters in enemy lands across the Danube, lest their requirements prove too great a strain on the agricultural and financial resources of the empire's provinces south of the river. Exasperated by this last demand, the soldiers rose in revolt, put a junior officer named Phocas at their head, and marched on Constantinople. Again becoming politically active, the Blues and Greens united against Maurice, and the aged emperor watched as his five sons were slaughtered before he himself met a barbarous death.

The ensuing reign of Phocas (602–610) may be described as a disaster. Khosrow seized the opportunity offered him by the murder of his benefactor, Maurice, to initiate a war of revenge that led Persian armies into the Anatolian heartland. Subsidies again failed to restrain the barbarians north of the Danube; after 602 the frontier crumbled, not to be restored save at the cost of centuries of warfare. Lacking a legitimate title, holding his crown only by right of conquest, Phocas found himself confronted by constant revolt and rebellion. To contemporaries, the coincidence of pestilence, endemic warfare, and social upheaval seemed to herald the coming of the Antichrist, the resurrection of the dead, and the end of the world.

But it was a human saviour who appeared, albeit under divine auspices. Heraclius, son of the Exarch of Africa, set sail from the western extremes of the empire, placing his fleet under the protection of an icon of the Virgin against Phocas, stigmatized in the sources as the "corrupter of virgins." In the course of his voyage along the northern shores of the Mediterranean, Heraclius added to his forces and arrived at Constantinople in October 610 to be hailed as a saviour. With the warm support of the Green faction, he quickly bested his enemy, decapitating Phocas and, with him, those Phocas had advanced to high civil and military office. There were, in consequence, few experienced counselors to aid Heraclius, for, among the men of prominence under Phocas—and earlier under Maurice—few survived to greet the new emperor.

THE 7TH CENTURY: THE HERACLIANS AND THE CHALLENGE OF ISLĀM

Heraclius and the origin of the themes. The most threatening problem Heraclius faced was the external menace of the Avars and the Persians, and neither people abated its pressure during the first years of the new reign. The Avars almost captured the Emperor in 617 during a conference outside the long walls protecting the capital. The Persians penetrated Asia Minor and then turned to the south, capturing Jerusalem and Alexandria (in Egypt). The great days of the Persian Achaemenid Empire seemed to have come again, and there was little in the recent history of the Byzantine emperors that would encourage Heraclius to place much faith in the future. He clearly could not hope to survive unless he kept under arms the troops he had brought with him; yet, the fate of Maurice demonstrated that this would be no easy task, given the empire's lack of financial and agricultural resources.

Three sources of Heraclius' strength

Three sources of strength enabled Heraclius to turn defeat into victory. The first was the pattern of military government as he and the nucleus of his army would have known it in the exarchates of North Africa or Ravenna. As it had been in the West, so it now was in the East. Civil problems were inseparable from the military: Heraclius could not hope to dispense justice, collect taxes, protect the church, and assure the future to his dynasty unless military power reinforced his orders. A system of military government, the exarchate, had accomplished these objectives so well in the West that, in a moment of despair, Heraclius sought to return to the land of his origins. In all likelihood, he applied similar principles of military rule to his possessions throughout Asia Minor, granting his generals (*stratēgoi*) both civil and military authority over those lands that they occupied with their "themes," as the army groups, or corps, were called in the first years of the 7th century.

Second, during the social upheaval of the previous decade, the imperial treasury had doubtless seized the estates of prominent individuals who had been executed either during Phocas' reign of terror or after his death. In consequence, though the treasury lacked money, it nonetheless possessed land in abundance, and Heraclius could easily have supported with grants of land those cavalry soldiers whose expenses in horses and armament he could not hope to meet with cash. If this hypothesis is correct, then, even before 622, themes, or army groups—including the guards (Opsikioi), the Armenians (Armeniakoi), and the Easterners (Anatolikoi)—were given lands and settled throughout Asia Minor in so permanent a fashion that, before the century was out, the lands occupied by these themes were identified by the names of those who occupied them. The Opsikioi were to be found in the Opsikion theme, the Armeniakoi in the Armeniakon, and the Anatolikoi in the Anatolikon. The term theme ceased thereafter to identify an army group and described instead the medieval Byzantine unit of local administration, the theme under the authority of the themal commander, the general (*stratēgos*).

When Heraclius "went out into the lands of the themes" in 622, thereby undertaking a struggle of seven years' duration against the Persians, he utilized the third of his sources of strength: religion. The warfare that ensued was nothing less than a holy war: it was partly financed by the treasure placed by the church at the disposal of the state; the Emperor's soldiers called upon God to aid them as they charged into battle; and they took comfort in the miraculous image of Christ that preceded them in their line of march. A brief summary of the campaign unfortunately gives no idea of the difficulties Heraclius encountered as he liberated Asia Minor (622); fought in Armenia with allies found among the Christian Caucasian peoples, the Lazi, the Abasgi, and the Iberians (624); and struggled in far-distant Lazica while Constantinople withstood a combined siege of Avars and Persians (626). An alliance with the Khazars, a Turkic people from north of the Caucasus, proved of material assistance in those years and of lasting import in Byzantine diplomacy. Heraclius finally destroyed the main Persian host at Nineveh in 627 and, after occupying Dastagird in 628, savoured the full flavour of triumph when his enemy, Khosrow, was deposed and murdered. The Byzantine emperor might well have believed that, if the earlier success of the Persians signalized the resurrection of the Achaemenid Empire, his own successes had realized the dreams of Caesar, Augustus, and Trajan.

Defeat of the Persians

Yet this was a war fought by medieval Byzantium and not by ancient Rome. Its spirit was manifest in 630, when Heraclius triumphantly restored the True Cross to Jerusalem, whence the Persians had stolen it, and—even more—when Constantinople resisted the Avar–Persian assault of 626. During the attack, the patriarch Sergius maintained the morale of the valiant garrison by proceeding about the walls, bearing the image of Christ to ward off fire, and by painting upon the gates of the western walls images of the Virgin and child to ward off attacks launched by the Avars—the "breed of darkness." The Avars withdrew when Byzantine ships defeated the canoes manned by Slavs, upon whom the nomad Avars depended for their naval strength. The latter never recovered from their defeat. As their empire crumbled, new peoples from the Black Sea to the Balkans emerged to seize power: the Bulgars of Kuvrat, the Slavs under Samo, and the Serbs and Croats whom Heraclius permitted to settle in the northwest Balkans once they had accepted Christianity.

As for the Byzantine defenders of Constantinople, they celebrated their victory by singing Romanos' great hymn "Akathistos," with choir and crowd alternating in the chant of the "Alleluia." The hymn, still sung in a Lenten service, commemorates those days when Constantinople survived as a fortress under ecclesiastical leadership, its defenders protected by the icons and united by their liturgy. This they sang in Greek, as befitted a people whose culture was now Greek and no longer Latin.

The successors of Heraclius: Islām and the Bulgars. In the same year that Heraclius went out into the themes, Muḥammad made his withdrawal (*hijrah*) from Mecca to Medina, where he established the *ummah,* or Muslim community. Upon the Prophet's death in 632, the caliphs, or successors, channeled the energies of the Arab Bedouin by launching them upon a purposive and organized plan of conquest. The results were spectacular: a Byzantine army was defeated at the Battle of the Yarmūk River (636), thereby opening Palestine and Syria to Arab Muslim control. Alexandria capitulated in 642, removing forever the province of Egypt from Byzantine authority. The Arabs had, meanwhile, advanced into Mesopotamia, capturing the royal city of Ctesiphon and, eventually, defeating an army under command of the Persian king himself. So ended the long history of Persia under Achaemenids, Parthians, and Sāsānians; further conquests were shortly to initiate that region's Islāmic phase (see further IRAN: *History;* ISLĀMIC WORLD, THE).

The rise of Islām

At least three aspects of the contemporary situation of Byzantium and Persia account for the phenomenal ease with which the Arabs overcame their enemies: first, both empires, exhausted by wars, had demobilized before 632; second, both had ceased to support those client states on the frontiers of the Arabian Peninsula that had restrained the Bedouin of the desert for a century past; third, and particularly in reference to Byzantium, religious controversy had weakened the loyalties that Syrians and Egyptians rendered to Constantinople. Heraclius had sought in 638 to placate Monophysite sentiment in these two provinces by promulgating the doctrine of Monothelitism, holding that Christ, although of two natures, had but one will. Neither in the East nor in the West did this compromise prove successful. The victorious Muslims granted religious freedom to the Christian community in Alexandria, for example, and the Alexandrians quickly recalled their exiled Monophysite patriarch to rule over them, subject only to the ultimate political authority of the conquerors. In such a fashion the city persisted as a religious community under an Arab Muslim domination more welcome and more tolerant than that of Byzantium.

The aging Heraclius was unequal to the task of containing this new menace, and it was left to his successors—Constantine III (ruled February to May 641), Constans II (641–668), Constantine IV (668–685), and Justinian II (685–695, 705–711)—to do so. This bare list of emperors obscures the family conflicts that often imperiled the succession, but gradually the principle was established that, even if brothers ruled as coemperors, the senior's authority would prevail. Although strife between Blues and Greens persisted throughout the century, internal revolt failed to imperil the dynasty until the reign of Justinian II. The latter was deposed and mutilated in 695. With the aid of the Bulgars, he returned in 705 to reassume rule and wreak a vengeance so terrible that his second deposition, and death, in 711 is surprising only in its delay of six years. From 711 until 717 the fortunes of the empire foundered; in that year, Leo, *stratēgos* of the Anatolikon theme, arrived as a second Heraclius to found a dynasty that would rescue the empire from its new enemies, the Arab Muslims and the Bulgars.

Three features distinguish the military history of the years 641–717: first, an increasing use of sea power on the part of the Arabs; second, a renewed threat in the Balkans occasioned by the appearance of the Onogur Huns, known in contemporary sources as the Bulgars; third, a persisting interest among the emperors in their western possessions, despite the gradual attrition of Byzantine authority in the exarchates of Carthage and Ravenna. Thanks to the control that the Arabs gradually asserted over the sea routes to Constantinople, they climaxed their earlier assaults on Armenia and Asia Minor with a four years' siege of the great city itself (674–678). Defeated in this last attempt by the use of Greek fire, a flammable liquid of uncertain composition, the Arabs signed a 30 years' truce, according to which they agreed to pay tribute in money, men, and horses. Lured by the unsettled conditions following Justinian's second deposition, they renewed their assaults by land and sea, and in 717 the Arabs were again besieging Constantinople.

The Bulgar menace

On the Balkan frontier, meanwhile, the Bulgars assumed the role abdicated by the Avars after 626. A pagan people whom the Khazars had forced toward the Danube Delta in the latter part of the 7th century, they eluded Constantine IV's attempts to defeat them in 681. By virtue of a treaty signed in that year, as well as others dating from 705 and 716, the Bulgars were recognized as an independent kingdom, occupying (to the humiliation of Byzantium) lands south of the Danube into the Thracian plain. While the Bulgars had thus deprived the empire of control in the north and central Balkans, the Byzantines could take comfort in the expeditions of 658 and 688/689 launched, respectively, by Constans II and Justinian II into Macedonia and in the formation of the themes of Thrace (687) and Hellas (695); these moves were evidence that Byzantine authority was beginning to prevail along the peninsular coastline and in certain parts of Greece where Slavs had penetrated.

In the West, the situation was less reassuring. Monothelitism had evoked a hostile reception among the churches of North Africa and Italy, and the resulting disaffection had encouraged the exarchs of both Carthage (646) and Ravenna (652) to revolt. By the end of the century, Africa had been largely lost to Muslim conquerors who would, in 711, seize the last outpost at Septem. For the moment Sicily and the scattered Italian possessions remained secure. Constans undertook operations against the Lombards, and he apparently intended to move his capital to Sicily, before his assassination ended the career of the last Eastern emperor to venture into the West. In summary, Leo III in 717 ruled over an empire humiliated by the presence of pagan barbarians upon Balkan soil rightfully considered "Roman," threatened by an attack upon its Anatolian heartland and its capital, and reduced, finally, in the West to Sicily and the remnants of the Ravenna exarchate.

Economics and institutions

However dismal the military record, institutional and economic developments had permitted the empire to survive and were to provide foundations for greater success in the centuries to come. The themal system had taken root and, with it, probably the institution of soldiers' properties. Military service was a hereditary occupation: the eldest son assumed the burden of service, supported primarily by revenues from other members of the family who worked the land in the villages. This last was a task easier to accomplish at the end of the 7th century thanks to the colonies of Slavs and other peoples brought into the empire and settled in the rural areas by Heraclius, Constantine IV, and Justinian II. In the 8th and 9th centuries, other emperors, including Leo III, Constantine V, and Nicephorus I, were to continue the practice, thus ending the population decline that had long eroded the ranks of Byzantine society. There are unmistakable signs of agricultural expansion even before 800; and, at about that time, urban life, which had never vanished in Asia Minor, began to flourish and expand in the Balkans. To judge from the evidence of the Farmer's Law, dated in the 7th century, the technological base of Byzantine society was more advanced than that of contemporary western Europe: iron tools could be found in the villages; water mills dotted the landscape; and field-sown beans provided a diet rich in protein. None of these advances was to characterize western European agriculture until the 10th century. Byzantine agriculture enjoyed the further advantage of a highly developed tradition of careful farming that persisted even in the darkest days, enabling the peasant to make the most of the soil upon which he worked. The invasions had even provided a form of stimulus to development: having lost first its Egyptian granary and, later, its North African and Sicilian resources, the empire had to live essentially, although not totally, from whatever it could produce in the lands remaining to it. The invasions had also, in all probability, broken up many a large estate, and the small peasant holding seems to have been the "normal" form of rural organization in this period. Although collective village organization persisted in the form of the rural commune and, with it, certain collective agricultural practices, the state seems to have made little or no attempt to bind the peasant to the soil upon which the tax registers had inscribed him. While Byzantium remained a slave-owning society, the *colonus* of the later Roman Empire had vanished, and a greater degree of freedom and mobility characterized agricultural relationships during the 7th and 8th centuries.

So it was, too, in trade and commerce. After the loss of Egypt and North Africa, the grain fleets manned by hereditary shipmasters disappeared; in their place there emerged the independent merchant, of sufficient importance to call forth a code of customary law, the Rhodian Sea Law, to regulate his practices. Military and religious hostilities failed to check him as he traded with the Bulgars in Thrace and, through Cyprus, with the Arabs. Despite constant warfare, this was, in short, a healthier society than the late Roman, and its chances of survival were further increased when the sixth general council (680–681) condemned Monothelitism and anathematized its adherents. With Egypt and Syria under Muslim rule, it was no longer necessary to placate Eastern Monophysitism, and it seemed that doctrinal discord would no longer separate Constantinople from the West. Events were to prove otherwise.

THE AGE OF ICONOCLASM: 717–867

The veneration of icons

For more than a century after the accession of Leo III (717–741), a persisting theme in Byzantine history may be found in the attempts made by the emperors, often with wide popular support, to eliminate the veneration of icons, a practice that had earlier played a major part in creating the morale essential to survival. The sentiment had grown in intensity during the 7th century; the Quinisext Council (Council in Trullo) of 692 had decreed that Christ should be represented in human form rather than, symbolically, as the lamb. The reigning emperor, Justinian II, had taken the unprecedented step of placing the image of Christ on his coinage while proclaiming himself the "slave of God." Evidence of a reaction against such iconodule (or image venerating) doctrines and practices may be found early in the 8th century, but full-fledged Iconoclasm (or destruction of the images) emerged as an imperial policy only when Leo III issued his decrees of 730. Under his son, Constantine V (ruled 741–775), the iconoclastic movement intensified, taking the form of violent persecution of the monastic clergy, the foremost defenders of the iconodule position. The Council of Nicaea in 787 restored iconodule doctrine at the instigation of the empress Irene, but military reversals led Leo V to resurrect in 815 the iconoclastic policies associated with Constantine V, one of Byzantium's most successful generals. Not until 843 were the icons definitively restored to their places of worship and icon veneration solemnly proclaimed as Orthodox belief. Even this brief summary suggests that the Emperor's fortunes on the battlefield were of no small moment in determining his attitude toward the icons, those channels whence superhuman power descended to man. An account of the age of Iconoclasm opens appropriately, then, with its military history.

The reigns of Leo III (the Isaurian) and Constantine V. Almost immediately upon Leo's accession, the empire's fortunes improved markedly. With the aid of the Bulgars, he turned back the Muslim assault in 718 and, in the intervals of warfare during the next 20 years, addressed himself to the task of reorganizing and consolidating the themes in Asia Minor. Thanks to the assistance of the traditional allies, the Khazars, Leo's reign concluded with a major victory, won again at the expense of the Arabs, at Acroenos (740). His successor, Constantine, had first to fight his way to the throne, suppressing a revolt of the Opsikion and Armeniakon themes launched by his brother-in-law Artavasdos. During the next few years, internal disorder in the Muslim world played into Constantine's hands as the ʿAbbāsid house fought to seize the caliphate from the Umayyads. With his enemy thus weakened, Constantine won noteworthy victories in northern Syria, transferring the prisoners he had captured there to Thrace in preparation for the wars against the Bulgars that were to occupy him from 756 to 775. In no fewer than nine campaigns, he undermined Bulgar strength so thoroughly that the northern enemy seemed permanently weakened, if not crushed. Even the venom used by the iconodule chroniclers of Constantine's reign cannot disguise the enormous popularity his victories won him.

In later centuries, the folk of Constantinople would stand by his tomb, seeking his aid against whatever enemy imperiled the city's defenses.

The empress Irene

Constantine's weak successors. His successors all but let slip the gains won by the great iconoclast. Constantine's son Leo IV died prematurely in 780, leaving to succeed him his 10-year-old son, Constantine VI, under the regency of the empress Irene. Not much can be said for Constantine, and Irene's policies as regent and (after the deposition and blinding of her son at her orders) as sole ruler from 797 to 802 were all but disastrous. Her iconodule policies alienated many among the themal troops, who were still loyal to the memory of the great warrior emperor, Constantine V. In an effort to maintain her popularity among the monkish defenders of the icons and with the population of Constantinople, she rebated taxes to which these groups were subject; she also reduced the customs duties levied outside the port of Constantinople, at Abydos and Hieros. The consequent loss to the treasury weighed all the more severely since victories won by the Arabs in Asia Minor (781) and by the Bulgars (792) led both peoples to demand and receive tribute as the price of peace. A revolt of the higher palace officials led to Irene's deposition in 802, and the so-called Isaurian dynasty of Leo III ended with her death, in exile, on the isle of Lesbos.

In the face of the Bulgar menace, none of the following three emperors succeeded in founding a dynasty. Nicephorus I (ruled 802–811), the able finance minister who succeeded Irene, reimposed the taxes that the Empress had remitted and instituted other reforms that provide some insight into the financial administration of the empire during the early 9th century. In the tradition of Constantine V, Nicephorus strengthened the fortifications of Thrace by settling, in that theme, colonists from Asia Minor.

Taking arms himself, he led his troops against the new and vigorous Bulgar khan, Krum, only to meet defeat and death at the latter's hands. His successor, Michael I Rhangabe (811–813), fared little better; internal dissensions broke up his army as it faced Krum near Adrianople, and the resulting defeat cost Michael his throne. In only one respect does he occupy an important place in the annals of the Byzantine Empire. The first emperor to bear a family name, Michael's use of the patronymic, Rhangabe, bears witness to the emergence of the great families, whose accumulation of landed properties would soon threaten the integrity of those smallholders upon whom the empire depended for its taxes and its military service. The name Rhangabe seems to be a Hellenized form of a Slav original (*rokavu*), and, if so, Michael's ethnic origin and that of his successor, Leo V the Armenian (ruled 813–820), provide evidence enough of the degree to which Byzantium in the 9th century had become not only a melting-pot society but, further, a society in which even the highest office lay open to the man with the wits and stamina to seize it. Leo fell victim to assassination, but before his death events beyond his control had improved the empire's situation. Krum died suddenly in 814 as he was preparing an attack upon Constantinople, and his son, Omortag, arranged a peace with the Byzantine Empire in order to protect the western frontiers of his Bulgar empire against the pressures exerted by Frankish expansion under Charlemagne and his successors. Since the death of the fifth caliph, Hārūn ar-Rashīd, had resulted in civil war in the Muslim world, hostilities from that quarter ceased. Leo used the breathing space to reconstruct those Thracian cities that the Bulgars had earlier destroyed. His work indicates the degree of gradual Byzantine penetration into the coastal fringes of the Balkan Peninsula, as does the number of themes organized in that same region during the early 9th century: those of Macedonia, Thessalonica, Dyrrhachium, Dalmatia, and the Strymon.

Michael II and Thomas the Slavonian

The new emperor, Michael II, was indeed able to establish a dynasty—the Amorian, or Phrygian—his son Theophilus (829–842) and his grandson Michael III (842–867) each occupying the throne in turn, but none would have forecast so happy a future during Michael II's first years. Thomas the Slavonian, Michael's former comrade in arms, gave himself out to be the unfortunate Constantine VI and secured his coronation at the hands of the Patriarch of Antioch; this was accomplished with the willing permission of the Muslim caliph under whose jurisdiction Antioch lay. Thomas thereupon marched to Constantinople at the head of a motley force of Caucasian peoples whose sole bonds were to be found in their devotion to iconodule doctrine and their hatred of Michael's Iconoclasm. Assisted by Omortag and relying upon the defenses of Constantinople, Michael defeated his enemy, but the episode suggests the tensions beneath the surface of Byzantine society: the social malaise, the ethnic hostility, and the persisting discord created by Iconoclasm. All these may explain the weakness displayed throughout Theophilus' reign, when a Muslim army defeated the Emperor himself (838) as a prelude to the capture of the fortress of Amorium in Asia Minor. It may also explain the concurrent decline of Byzantine strength in the Mediterranean, manifest in the capture of Crete by the Arabs (826 or 827) and in the initiation of attacks upon Sicily that finally secured the island for the world of Islām. Iconoclasm certainly played its part in the further alienation of East from West, and a closer examination of its doctrines will suggest why this may have been.

The Iconoclastic Controversy. Iconoclasts and iconodules agreed on one fundamental point: a Christian people could not prosper unless it assumed the right attitude toward the holy images, or icons. They disagreed, of course, on what that attitude should be. Each could discover supporting arguments in the writings of the early church, and it is essential to remember that the debate over images is as old as Christian art. The fundamentals of Iconoclasm were by no means an 8th-century discovery. The ablest defender of the iconodule position was, however, the 8th-century theologian St. John of Damascus. Drawing upon Neoplatonic doctrine, John suggested that the image was but a symbol; the creation of the icon was justified since, by virtue of the Incarnation, God had himself become man.

The iconoclasts responded by pointing to the express wording of the Second Commandment. The condemnation therein of idolatry seems to have weighed heavily with Leo III, who may have been influenced by Islām, a religion that strictly prohibited the use of religious images. The latter point is debatable, as is the contention that Iconoclasm was particularly an expression of sentiment to be found in the eastern themes of the empire. There is little doubt, however, that Monophysitism influenced the ideas of Constantine V and, through him, the course of debate during the last half of the 8th century. In the eyes of the Monophysite, who believed in the single, indistinguishable, divine nature of Christ, the iconodule was guilty of sacrilege. Either he was a Nestorian, reducing the divine nature to human terms in the image, or he was a Chalcedonian Dyophysite, radically distinguishing that which man could not distinguish. Still another consideration favouring Iconoclasm may be found in the intimate connection of iconoclastic doctrine with the emperor's conception of his role as God's vicegerent on Earth. During the late 6th and 7th centuries, iconodule emperors had viewed themselves in a pietistic fashion, emphasizing their devotion and subservience to God. Constantine V, on the other hand, pridefully replaced the icons with imperial portraits and with representations of his own victories. Viewed in this light, Iconoclasm signaled a rebirth of imperial confidence; so deservedly great was Constantine's reputation, and so dismal were the accomplishments of his successors, that a Leo V, for one, could easily believe that God favoured the iconoclastic battalions.

Under Constantine V, the struggle against the icons became a struggle against their chief defenders, the monastic community. The immediate destruction wrought by Constantine and his zealous subordinates is, however, of less moment than the lasting effect of the persecution on the Orthodox clergy. Briefly put, the church became an institution rent by factions, wherein popular discontent found a means of expression. Intransigent iconodules looked for their leaders among the monks of Studion, the monastery founded by Studius, and they found one in the person of the monastery's abbot, St. Theodore Studites (759–826). In the patriarch Ignatius (847–858; 867–877) they discovered a spokesman after their own hearts: one drawn from the monastic ranks and contemptuous of all the allurements that the world of secular learning seemed to offer. More significant than the men to be found on the other extreme, iconoclast patriarchs, including Anastasius and John Grammaticus, were the representatives of the moderate party, composed of the patriarchs Tarasius, Nicephorus, Methodius, and Photius. Although iconodule in sympathy, the group enjoyed little rapport with the monastic zealots. Unlike the average monk, they were often educated laymen, trained in the imperial service and ready to compromise with imperial authority.

Byzantine relations with Catholic Europe

Not only was Iconoclasm a major episode in the history of the Byzantine, or Orthodox, Church, but it also permanently affected relations between the empire and Roman Catholic Europe. The Lombard advance, it may be remembered, had restricted Byzantine authority in Italy to the Exarchate of Ravenna, and to that quarter the popes of the 7th century, themselves ordinarily of Greek or Syrian origin, turned for protection against the common enemy. During the 8th century, two issues alienated Rome from Constantinople: Iconoclasm and quarrels stemming from the question of who should enjoy ecclesiastical jurisdiction over Illyricum and over Calabria in southern Italy. Pope Gregory II refused to accept the iconoclastic doctrines of Leo III; and his successor, Gregory III, had openly to condemn them at a council. Once Ravenna fell to the Lombards, and the exarchate ceased to exist in 751, the papacy had to seek a new protector. This was found in the person of the Frankish leader Pepin III, who sought some form of sanction to legitimize his seizure of the crown from the feeble hands of the last representative of the Merovingian dynasty. Thus Pope Stephen II (or III) anointed Pepin as king of the Franks in 754, and the latter entered Italy to take arms against the Lombard king. Even the restoration of icon veneration in 787 failed to bridge the differences between Orthodox Byzantium and Catholic Europe, for the advisers of Pepin's son and successor, Charlemagne, condemned the iconodule position as heartily as an earlier generation had rejected the iconoclast decrees of Leo III. Nor could the men of Charlemagne's time admit that a woman—the empress Irene—might properly assume the dignity of emperor of the Romans. For all these reasons, Charlemagne, king of the Franks and Lombards by right of conquest, assented to his coronation as emperor of the Romans on Christmas Day, 800, by Pope Leo III. No longer a barbarian king, Charlemagne became, by virtue of the symbolism of the age, a new Constantine. This the Byzantine chancery could not accept, for, if there were one God, one faith, and one truth, then there could be but one empire and one emperor; surely that emperor ruled in Constantinople, not in Charlemagne's Aachen. Subsequent disputes between Rome and Constantinople seemed often to centre upon matters of ecclesiastical discipline; underlying these differences were two more powerful considerations, neither of which could be ignored. According to theory there could be but one empire; clearly, there were two. And between Rome and Constantinople there stood two groups of peoples open to conversion: the Slavs of central Europe and the Bulgars in the Balkans. From which of the two jurisdictions would these people accept their Christian discipline? To which, in consequence, would they owe their spiritual allegiance?

The reign of Michael III

The reign of Michael III (ruled 842–867) draws together these and other threads from the past. Veneration of the icons was definitely rehabilitated in 843; and it was done in so diplomatic a fashion that the restoration, in itself, produced no new rifts, although old factionalisms persisted with the appointment of a monk, Ignatius, as patriarch. The latter's intransigent zealotry found little favour with Caesar Bardas, Michael's uncle, who had seized power from the Empress Regent in 856. Two years later Ignatius was deposed and replaced by a moderate: the scholar and layman Photius. No single person better exemplified the new age, nor, indeed, did any other play a larger part in the cultural rebirth and missionary activity among the Slavs, Bulgars, and Russians, which mark the middle of the 9th century. The same aggressive and enterprising spirit is manifest in the military successes won on the Asia Minor frontier, culminating in Petronas' victory at Poson (863) over the Muslim emir of Melitene.

In Sicily, and throughout the Mediterranean, Byzantine arms were less successful, but, thanks to Photius' diplomatic skill, the see of Constantinople maintained its position against Rome during the so-called Photian Schism. When Pope Nicholas I challenged Photius' elevation to the patriarchate, deploring as uncanonical the six days' speed with which he had been advanced through the successive ranks of the hierarchy, the Byzantine patriarch refused to bow. He skillfully persuaded Nicholas' delegates to a council summoned at Constantinople to investigate the matter that he was the lawful patriarch despite the persisting claims of the rival Ignatian faction. Nicholas, alleging that his men had been bribed, excommunicated Photius; a council at Constantinople responded (867) by excommunicating Nicholas in turn. The immediate issues between the two sees were matters of ecclesiastical supremacy, the liturgy, and clerical discipline; behind these sources of division lay the question of jurisdiction over the converts in Bulgaria. And behind that question may be found centuries of growing separation between the minds and institutions of the eastern and the western Mediterranean worlds, symbolized in the roles assumed by two among the major protagonists in the Photian Schism. It was the supreme spiritual authority, the pope, who hurled anathemas from the West, but it was God's vicegerent on Earth, the emperor Michael III, who presided at the council of 867.

Michael did not long survive this moment of triumph. Later that year, he was murdered by his favourite, Basil, who, on his bloody path to the throne, had earlier disposed of Caesar Bardas. As had Heraclius and Leo III before him, Basil came to found a dynasty, in this instance the Macedonian house. Unlike his predecessors, he came not as a saviour but as a peasant adventurer to seize an already sound empire whose next centuries were to be its greatest (see also EASTERN ORTHODOXY: *History*).

(J.L.Te./D.M.N.)

From 867 to the Ottoman conquest

THE MACEDONIAN ERA: 867–1025

Under the Macedonians, at least until the death of Basil II in 1025, the empire enjoyed a golden age. Its armies regained the initiative against the Arabs in the East, and its missionaries evangelized the Slavs, extending Byzantine influence in Russia and the Balkans. And, despite the rough military character of many of the emperors, there was a renaissance in Byzantine letters and important developments in law and administration. At the same time there were signs of decay: resources were squandered at an alarming rate; there was growing estrangement from the West; and a social revolution in Anatolia was to undermine the economic and military strength of the empire.

Founding of the dynasty

The empire was in theory an elective monarchy with no law of succession. But the desire to found and perpetuate a dynasty was strong, and it was often encouraged by popular sentiment. This was especially true in relation to the Macedonian dynasty, the founder, Basil I, having murdered his way to the throne in 867. Probably of Armenian descent, though they had settled in Macedonia, Basil's family was far from distinguished and can hardly have expected to produce a line of emperors that lasted through six generations and 189 years. But, having acquired the imperial crown, Basil tried to make sure that his family would not lose it and nominated three of his sons as coemperors. Though he was his least favourite,

The Macedonian Emperors, 867–1025

name	reign dates
Basil I	867–886
Leo VI	886–912
Constantine VII	913–959
Romanus I*	919–944
Romanus II	959–963
Nicephorus II	963–969
John I	969–976
Basil II	976–1025

*Ruled in Constantine's name.

through the scholarly Leo VI, who succeeded him in 886, the succession was at least secure. Even the three soldier-emperors who usurped the throne during the Macedonian era were conscious, in varying degrees, that they were protecting the rights of a legitimate heir during a minority: Romanus I Lecapenus for Constantine VII, the son of Leo VI; and Nicephorus Phocas and John Tzimisces for Basil II, the grandson of Constantine VII.

Military revival. A reassertion of Byzantine military and naval power in the East began with victories over the Arabs by Michael III's general Petronas in 856. From 863 the initiative lay with the Byzantines. The struggle with the Arabs, which had long been a struggle for survival, became a mounting offensive that reached its brilliant climax in the 10th century. By 867 a well-defined boundary existed between the Byzantine Empire and the territory of the ʿAbbāsid caliphate. Its weakest point was in the Taurus Mountains above Syria and Antioch. Basil I directed his operations against this point, recovered Cyprus for a while, and campaigned against the Paulicians, a heretical Christian sect whose anti-imperial propaganda was effective in Anatolia. But the conflict with Islām was one that concerned the whole empire, in the West as well as in the East, and by sea as well as by land. In 902 the Arabs completed the conquest of Sicily, but they were kept out of the Byzantine province of South Italy, for whose defense Basil I had even made some effort to cooperate with the Western emperor Louis II. The worst damage, however, was done by Arab pirates who had taken over the island of Crete. In 904 they plundered Thessalonica, carrying off quantities of loot and prisoners. Leo VI sent a naval expedition to Crete in 911, but the Muslims drove it off and humiliated the Byzantine navy off Chios in 912.

Success of eastern offensive

On the eastern frontier, the Byzantine offensive was sustained with great success during the reign of Romanus I Lecapenus by an Armenian general John Curcuas (Gurgen), who captured Melitene (934) and then Edessa (943), advancing across the Euphrates into the caliph's territory. It was Curcuas who paved the way for the campaigns of the two soldier-emperors of the next generation. In 961 Nicephorus Phocas, then domestic (commander) of the armies in the West, reconquered Crete and destroyed the Arab fleet that had terrorized the Aegean for 150 years; he thereby restored Byzantine naval supremacy in the eastern Mediterranean. In 962 his strategy achieved unexpected triumphs all along the eastern frontier and culminated in the capture of Aleppo in Syria. When he was proclaimed emperor in March 963, Nicephorus appointed another Armenian general, John Tzimisces, as domestic of the East, though he retained personal command of operations against the Arabs. By 965 he had driven them out of Cyprus and was poised for the reconquest of Syria. The revived morale and confidence of Byzantium in the East showed itself in the crusading zeal of Nicephorus Phocas and John Tzimisces for the reconquest of Syria and the Holy Land. The ground lost to Islām in the 7th century was thus fast being regained; and, although Jerusalem was never reached, the important Christian city of Antioch, seat of one of the patriarchs, was recaptured in 969. These victories were achieved largely by the new cavalry force built up by Nicephorus Phocas. In the areas recovered from the Arabs, land was distributed in military holdings with the interests of the cavalry in mind. But the victories were achieved at the expense of the western provinces, and an attempt to recover Sicily ended in failure in 965.

The campaigns of John Tzimisces, who usurped the throne in 969, were directed against the Emir of Mosul on the Tigris and against the new Fāṭimid caliph of Egypt, who had designs on Syria. By 975 almost all of Syria and Palestine, from Caesarea to Antioch, as well as a large part of Mesopotamia far to the east of the Euphrates, was in Byzantine control. The way seemed open for Tzimisces to advance to the ʿAbbāsid capital of Baghdad on the one hand and to Jerusalem and Egypt on the other. But he died in 976 and his successor, Basil II, the legitimate heir of the Macedonian house, concentrated most of his resources on overcoming the Bulgars in Europe, though he did not abandon the idea of further reconquest in the East. The Kingdom of Georgia (Iberia) was incorporated into the empire by treaty. Part of Armenia was annexed, with the rest of it to pass to Byzantium on the death of its king. Basil II personally led two punitive expeditions against the Fāṭimids in Syria, but otherwise his eastern policy was to hold and to consolidate what had already been gained. The gains can be measured by the number of new themes (provinces) created by the early 11th century in the area between Vaspurakan in the Caucasus and Antioch in Syria. The annexation of Armenia, the homeland of many of the great Byzantine emperors and soldiers, helped to solidify the eastern wall of the Byzantine Empire for nearly a century.

Relations with the Slavs and Bulgars. Although imperial territory in the East could be reclaimed only by military conquest, in the Balkans and in Greece the work of reclamation could be assisted by the diplomatic weapon of evangelization. The Slavs and the Bulgars could be brought within the Byzantine orbit by conversion to Christianity. The conversion of the Slavs was instigated by the patriarch Photius and carried out by the monks Cyril and Methodius from Thessalonica. Their invention of the Slavonic alphabet (Cyrillic and Glagolitic) made possible the translation of the Bible and the Greek liturgy and brought literacy as well as the Christian faith to the Slavic peoples. The work began in the Slavic Kingdom of Moravia and spread to Serbia and Bulgaria. Latin missionaries resented what they considered to be Byzantine interference among the northern Slavs, and there were repeated clashes of interest that further damaged relations between the sees of Rome and Constantinople. The conversion of the Bulgars became a competition between the two churches and was ably exploited by the Bulgar king Boris until, in 870, he opted for Orthodox Christianity on condition of having an archbishop of his own.

Conversion of the Slavs

Bulgarian wars. The trade with Constantinople that followed the missionaries whetted the appetites of the Slavs and Bulgars for a larger share in the material wealth of Byzantium. Simeon (Symeon) I of Bulgaria, who succeeded his father Boris in 893 and who had been educated at Constantinople, proved to be an even more dangerous enemy than the Arabs. His efforts to become emperor dominated Byzantine history for some 15 years. In 913 he brought his army to the walls of Constantinople, demanding the imperial title. The patriarch, Nicholas Mysticus, appeased Simeon for a time, but it was Romanus Lecapenus who, by patience and diplomacy, undermined the power of the Bulgars and thwarted Simeon's ambitions. Simeon died in 927, and his son Peter I came to terms with Byzantium and married a granddaughter of Romanus.

Relations with Russia. The Russians lay far outside the Roman jurisdiction. Their warships, sailing down the Dnepr from Kiev to the Black Sea, first attacked Constantinople in 860. They were beaten off, and almost at once Byzantine missionaries were sent into Russia. The Russians were granted trading rights in Constantinople in 911, but in 941 and 944, led by Prince Igor, they returned to the attack. Both assaults were repelled, and Romanus I set about breaking down the hostility and isolationism of the Russians by diplomatic and commercial contacts. In 957 Igor's widow, Olga, was baptized and paid a state visit to Constantinople during the reign of Constantine VII; her influence enabled Byzantine missionaries to work with greater security in Russia, thus spreading Christianity and Byzantine culture. Olga's son Svyatoslav was pleased to serve the empire as an ally against the Bulgars from

968 to 969, though his ambition to occupy Bulgaria led to war with Byzantium in which he was defeated and killed. In 971 John Tzimisces accomplished the double feat of humiliating the Russians and reducing Bulgaria to the status of a client kingdom. Byzantine influence over Russia reached its climax when Vladimir of Kiev, who had helped Basil II to gain his throne, received as his reward the hand of the Emperor's sister in marriage and was baptized in 989. The mass conversion of the Russian people followed, with the establishment of an official Russian Church subordinate to the patriarch of Constantinople.

Bulgar revolt. The Bulgars, however, were not content to be vassals of Byzantium and rebelled under Samuel, youngest of the four sons of a provincial governor in Macedonia. Samuel made his capital at Ochrida and created a Bulgarian empire stretching from the Adriatic to the Black Sea and even, for a while, into Greece, though Thessalonica remained Byzantine. The final settlement of the Bulgar problem was worked out by Basil II in a ruthless and methodical military campaign lasting for some 20 years, until, by 1018, the last resistance was crushed. Samuel's dominions became an integral part of the Byzantine Empire and were divided into three new themes. At the same time the Slav principalities of Serbia (Rascia and Dioclea) and Croatia became vassal states of Byzantium, and the Adriatic port of Dyrrhachium came under Byzantine control. Not since the days of Justinian had the empire covered so much European territory. But the annexation of Bulgaria meant that the Danube was now the only line of defense against the more northerly tribes, such as the Pechenegs, Cumans, and Magyars.

Bulgars crushed by Basil II

Estrangement from the West. The extension of Byzantine interests to the Adriatic, furthermore, had raised again the question of Byzantine claims to South Italy and, indeed, to the whole western part of the old Roman Empire. The physical separation of that empire into East and West had been emphasized by the settlement of the Slavs in the Balkan Peninsula and in Greece, and since the 7th century the two worlds had developed in their different ways. Their differences had been manifested in ecclesiastical conflicts, such as the Photian Schism. The conversion of the Slavs had produced bitterness between the agents of the rival jurisdictions. But the reestablishment of Byzantine authority in Greece and eastern Europe, added to the gains against the Muslim powers in Asia, reinforced the Byzantine belief in the universality of the empire, to which Italy and the West must surely be reunited in time. Until that time came, the fiction was maintained that the rulers of western Europe, like those of the Slavs, held their authority by virtue of their special relationship with the one true emperor in Constantinople.

It was sometimes suggested that a marriage alliance might bring together the Eastern and Western parts of the empire and so provide for a united defense against the common enemy in Sicily—the Arabs. In 944 Romanus II, son of Constantine VII, married a daughter of Hugh of Provence, the Carolingian claimant to Italy. Constantine VII also kept up diplomatic contact with Otto I, the Saxon king of Germany. But the case was dramatically altered when Otto was crowned emperor of the Romans in 962, for this was a direct affront to the unique position of the Byzantine emperor. Otto tried, and failed, to establish his claim, either by force in the Byzantine province in Italy or by negotiation in Constantinople. His ambassador Liudprand of Cremona wrote an account of his mission to Nicephorus Phocas in 968 and of the Emperor's scornful rejection of a proposed marriage between Otto's son and a Byzantine princess. The incident vividly demonstrates the superior attitude of the Byzantines toward the West in the 10th century. John Tzimisces relented to the extent of arranging for one of his own relatives to marry Otto II in 972, though the arrangement implied no recognition of a Western claim to the empire. Basil II agreed that Otto III also should marry a Byzantine princess. But this union was never achieved; and subsequently Basil reorganized the administration of Byzantine Italy and was preparing another campaign against the Arabs in Sicily at the time of his death in 1025. The myth of the universal Roman Empire died hard.

Holy Roman Empire founded in the West

Culture and administration. The Iconoclastic Controversy had aggravated the estrangement of the Byzantine Church and Empire from the West. But it helped to define the tenets of Orthodoxy; and it had an effect on the character of Byzantine society for the future. On the one hand, the church acquired a new unity and vitality: its missionaries spread the Orthodox faith in new quarters of the world, its monasteries proliferated, and its spiritual tradition was carried forward by the sermons and writings of the patriarch Photius in the 9th century and of Symeon the New Theologian in the 10th and 11th centuries. On the other hand, the empire became more aware of its Greco-Roman heritage. Interest in classical Greek scholarship revived following the reorganization of the University of Constantinople under Michael III. The revival was fostered and patronized particularly by the scholar-emperor Constantine VII Porphyrogenitus, who saw to the compilation of three great works on the administration, the court ceremonies, and the provinces of his empire. He also commissioned a history of the age to which he contributed a biography of his grandfather Basil I. The age produced little original research, but lexicons (such as the 10th-century *Suda*), anthologies, encyclopaedias, and commentaries (such as the *Lexicon* and *Bibliotheca* of Photius) were produced in great number. The soldier-emperors of the 10th century were less interested in intellectual pursuits, but scholarship received a new impetus in the 11th century with Michael Psellus.

The founder of the dynasty, Basil I, and his son Leo VI, made plain their intention to inaugurate a new era by a restatement of the imperial law. Basil hoped to make a complete revision of the legal code, but only a preliminary textbook (Procheiron) with an introduction (Epanagoge) appeared during his reign. Leo VI, however, accomplished the work with the publication of the 60 books of the Basilica, which Hellenized the legal code of Justinian and made it more intelligible and accessible to lawyers. Additions and corrections to meet the needs of the time were incorporated in Leo's 113 novels (decrees), which represent the last substantial reform of the civil law in Byzantium. Enshrined in this legislation was the principle of the absolute autocracy of the emperor as being himself the law. The Senate, the last vestige of Roman republican institutions, was abolished. Only in the matter of the spiritual welfare of his subjects did the emperor recognize any limits to his authority. The ideal relationship of a dyarchy between emperor and patriarch, the body and the soul of the empire, was written into the Epanagoge of Basil I, in a section probably composed by Photius.

Growth of bureaucracy

The administration in this period was ever more closely centralized in Constantinople, with an increasingly complex and numerous bureaucracy of officials who received their appointments and their salaries from the emperor. The emperor also controlled the elaborate machinery of the foreign and diplomatic service. Some of his civil servants, however, were powerful enough to play the part of kingmakers, notably Basil, the chamberlain who engineered the ascent to the throne of Nicephorus Phocas and John Tzimisces. Order and the regulation of trade, commerce, and industry in the capital were in the hands of the prefect of the city, whose functions are outlined in the 9th-century *Book of the Eparch.* He was responsible for organizing and controlling the guilds or colleges of craftsmen and retailers, whose legal rights and duties to the state were strictly circumscribed and supervised. The provinces in Europe and Asia were administered according to their territorial division into themes, which, by the 10th century, numbered more than 30. The themes, though subdivided and reduced in size, retained their military character. Their governors, or *stratēgoi,* combined military and civil authority and were directly answerable to the emperor, who appointed them. The army and the navy were, for the most part, recruited from the ranks of soldier-farmers who held hereditary grants of land within the territory of each theme. The border districts were protected by contingents of frontier troops, led by their own officers or lords of the marches. Their exploits and adventures were romanticized in the 10th-century folk-epic of Digenis Akritas. But warfare was studied and perfected as a science, and it was the subject of treatises such as the *Tactica* of Leo VI, derived from the *Strategicon* of the emperor Maurice.

Growth of landholding aristocracy

Social and economic change. The wars of reconquest on the eastern frontier in this period and the general military orientation of imperial policy brought to the fore a new class of aristocracy, whose wealth and power were based on land ownership and who held most of the higher military posts. Trade and industry in the cities were so rigidly controlled by the government that almost the only profitable form of investment for private enterprise was the acquisition of landed property. The military aristocracy, therefore, took to buying up the farms of free peasants and soldiers and reducing their owners to varying forms of dependence. As the empire grew stronger, the rich became richer. Given the system of agriculture prevailing in Anatolia and the Balkans, every failure of crops, every famine, drought, or plague produced a quota of destitute peasant-soldiers willing to turn themselves and their land over to the protection of a prosperous and ambitious landlord. The first emperor to see the danger in this development was Romanus I Lecapenus, who, in 922 and 934, passed laws to defend the small landowners against the acquisitive instincts of the "powerful"; for he realized that the economic as well as the military strength of the empire depended on the maintenance within the theme system of the institution of free, yet tax-paying, soldier-farmers and peasants in village communities. (Only freemen owed military service.)

Successive emperors after Romanus I enforced and extended his agrarian legislation. But the cost of the campaigns of reconquest from the Arabs had to be met by higher taxation, which drove many of the poorer peasants to sell their lands and to seek security as tenant farmers. Nicephorus Phocas, who belonged to one of the aristocratic landowning families of Anatolia, was naturally reluctant to act against members of his own class, though he adhered to the principle that the rights of the poor should be safeguarded. His laws about land tenure were particularly directed toward the creation of a more mobile force of heavy-armed cavalry recruited from those who could afford the equipment, which inevitably brought changes in the social structure of the peasant militia. On the other hand, Nicephorus took a firm line to prevent the accumulation of further land by the church, and he forbade any addition to the number of monasteries, whose estates, already extensive, were unproductive to the economy.

Basil II's land reforms

The last emperor to attempt to deal with the problem of land ownership seriously was Basil II, whose rise to the throne had involved the empire in a bitter and costly war against the aristocratic Sclerus and Phocas families. In 996 Basil promulgated comprehensive punitive legislation against the landed families, ordering the restitution of land acquired from the peasantry since 922 and requiring proof of title to other land going back in some cases as far as 1,000 years. Further, the system of collective responsibility for the payment of outstanding taxes known as the *allelengyon* now devolved not on the rest of the village community but on the nearest large landowner, whether lay or ecclesiastical. Basil's conquest of Bulgaria somewhat altered the social and economic pattern of the empire, for new themes were created there in which there was no long tradition of a landed aristocracy as in Anatolia. After his death in 1025 the powerful hit back, and the government in Constantinople was no longer able to check the absorption of small freeholders by the great landowners and the consequent feudalization of the empire.

This process was particularly disastrous for the military establishment. The success and prestige of the Byzantine Empire in the Macedonian era to a large extent depended on the unrivaled efficiency of its army in Anatolia. A professional force, yet mainly native to the soil and so directly concerned with the defense of that soil, it had no equal in the Western or the Arab world at the time. And yet it was in this institution that the seeds of decay and disintegration took root; for most of the army's commanders were drawn from the great landowners of Anatolia, who had acquired their riches and their status by undermining the social and economic structure on which its recruitment depended. Basil II had restrained them with such an iron hand that a reaction was inevitable after his death. Indeed, it is doubtful if Byzantine society could have tolerated another Basil II, despite all his triumphs. Soured by long years of civil war at the start of his reign, ascetic and uncultured by nature, Basil embodied the least attractive features of Byzantine autocracy. Some have called him the greatest of all the emperors. But the virtue of philanthropy, which the Byzantines prized and commended in their rulers, was not a part of his greatness; and the qualities that lent refinement to the Byzantine character, among them a love of learning and the arts, were not fostered during his reign. Yet, while Basil was busily earning his title of Bulgaroctonus ("Bulgar Slayer"), St. Symeon the New Theologian was exploring the love of God for man in some of the most poetic homilies in all mystical literature.

BYZANTINE DECLINE AND SUBJECTION TO WESTERN INFLUENCES: 1025–1260

Basil II never married. But after his death his relatives remained in possession of the throne until 1056, less because of their efficiency than because of a general feeling

The Byzantine Empire in 1025.

The later Macedonians

among the Byzantine people that the prosperity of the empire was connected with the continuity of the Macedonian dynasty. When Basil's brother Constantine VIII died in 1028, the line was continued in his two daughters, Zoe and Theodora. Zoe was married three times: to Romanus III Argyrus (ruled 1028–34), to Michael IV (1034–41), and to Constantine IX Monomachus (1042–55), who outlived her. When Constantine IX died in 1055, Zoe's sister, Theodora, reigned alone as empress until her death a year later.

The great emperors of the golden age, not all of them members of the Macedonian family, molded the history of that age. The successors of Basil II were rather the creatures of circumstances, because they did not make and seldom molded. In the 56 years from 1025 to 1081, there were 13 emperors. An attempt made by Constantine X Ducas to found a new dynasty was disastrously unsuccessful. Not until the rise of Alexius I Comnenus to power, in 1081, was stability restored by an ensured succession in the Comnenus family, who ruled for more than 100 years (1081–1185).

11th-century weakness. The state of the Byzantine Empire in the 11th century may be compared to that of the Roman Empire in the 3rd century, when, after a long period of secure prosperity, new pressures from beyond the frontiers aggravated the latent tensions in society. The brief reigns of Basil II's heirs reflected, and were often the product of, a division in the Byzantine ruling class, a conflict between the military aristocracy of the provinces and the civilian aristocracy, or bureaucracy, of Constantinople. Each faction put up rival emperors. The sophisticated urban aristocracy favoured rulers who would reverse the militaristic trend of the empire and who would expand the civil service and supply them and their families with lucrative offices and decorative titles. The military families, whose wealth lay not in the capital but in the provinces and who had been penalized by Basil II's legislation, favoured emperors who were soldiers and not civil servants. In this they were more realistic, for in the latter part of the 11th century it became ever clearer that the empire's military strength was no longer sufficient to hold back its enemies. The landowners in the provinces appreciated the dangers more readily than the government in Constantinople, and they made those dangers an excuse to enlarge their estates in defiance of all the laws passed in the 10th century. The theme system in Anatolia, which had been the basis of the empire's defensive and offensive power, was rapidly breaking down at the very moment when its new enemies were gathering their strength.

On the other hand, the urban aristocracy of Constantinople, reacting against the brutalization of war, strove to make the city a centre of culture and sophistication. The university was endowed with a new charter by Constantine IX in 1045, partly to ensure a steady flow of educated civil servants for the bureaucracy. The law school was revived under the jurist John Xiphilinus; the school of philosophy was chaired by Michael Psellus, whose researches into every field of knowledge earned him a reputation for omniscience and a great following of brilliant pupils. Psellus—courtier, statesman, philosopher, and historian—is in himself an advertisement for the liveliness of Byzantine society in the 11th century. What he and others like him failed to take into account was that their empire was more and more expending the resources and living on the reputation built up by the Macedonian emperors.

Arrival of new enemies. The new enemies that emerged in the 11th century, unlike the Arabs or the Bulgars, had no cause to respect that reputation. They appeared almost simultaneously on the northern, the eastern, and the western frontiers. It was nothing new for the Byzantines to have to fight on two fronts at once. But the task required a soldier on the throne. The Pechenegs, a Turkic tribe, had long been known as the northern neighbours of the Bulgars. Constantine VII had thought them to be valuable allies against the Bulgars, Magyars, and Russians. But after the conquest of Bulgaria, the Pechenegs began to raid across the Danube into what was then Byzantine territory. Constantine IX allowed them to settle south of the river, where their numbers and their ambitions increased. By the mid-11th century they were a constant menace to the peace in Thrace and Macedonia, and they encouraged the spirit of revolt among the Bogomil heretics in Bulgaria. It was left to Alexius I to avert a crisis by defeating the Pechenegs in battle in 1091.

Seljuq Turks in the East

The new arrivals on the eastern frontier were the Seljuq Turks, whose conquests were to change the whole shape of the Muslim and Byzantine worlds. In 1055, having conquered Persia, they entered Baghdad, and their prince assumed the title of sultan and protector of the ʿAbbāsid caliphate. Before long they asserted their authority to the borders of Fāṭimid Egypt and Byzantine Anatolia. They made their first explorations across the Byzantine frontier into Armenia in 1065 and, in 1067, as far west as Caesarea in central Anatolia. The raiders were inspired by the Muslim idea of holy war, and there was at first nothing systematic about their invasion. They found it surprisingly easy, however, to plunder the countryside and isolate the cities, owing to the long neglect of the eastern frontier defenses by the emperors in Constantinople. The emergency lent weight to the military aristocracy in Anatolia who, in 1068, finally secured the election of one of their own number, Romanus IV Diogenes, as emperor. Romanus assembled an army to deal with what he saw as a large-scale military operation. It was a sign of the times that his army was mainly composed of foreign mercenaries. In August 1071 it was defeated at Manzikert, near Lake Van in Armenia. Romanus was taken prisoner by the Seljuq sultan, Alp-Arslan. He was allowed to buy his freedom after signing a treaty, but the opposition in Constantinople refused to have him back as emperor and installed their own candidate, Michael VII. Romanus was treacherously blinded. The Seljuqs were thus justified in continuing their raids and were even encouraged to do so. Michael VII invited Alp-Arslan to help him against his rivals, Nicephorus Bryennius and Nicephorus Botaneiates, each of whom proclaimed himself emperor at Adrianople in 1077 and at Nicaea in 1078. In the four years of ensuing civil war there were no troops to defend the eastern frontier. By 1081 the Turks had reached Nicaea. The heart of the empire's military and economic strength, which the Arabs had never mastered, was now under Turkish rule.

Battle of Manzikert

The new enemies in the West were the Normans, who began their conquest of South Italy early in the 11th century. Basil II's project of recovering Sicily from the Arabs had been almost realized in 1042 by the one great general of the post-Macedonian era, George Maniaces, who was recalled by Constantine IX and killed as a pretender to the throne. The Normans thereafter made steady progress in Italy. Led by Robert Guiscard, they carried all before them; in April 1071, Bari, the last remaining Byzantine stronghold, fell after a three-year siege. Byzantine rule in Italy and the hope of a reconquest of Sicily were at an end.

The disasters at Manzikert and at Bari, in the same year 1071, at opposite extremes of the empire, graphically illustrate the decline of Byzantine power. The final loss of Italy seemed to underline the fact of the permanent division between the Greek East and the Latin West, which was now not only geographical and political but also increasingly cultural and ecclesiastical. In 1054 a state of schism had been declared between the churches of Rome and Constantinople. The political context of the event was the Norman invasion of Italy, which at the time was a matter of as much concern to the papacy as it was to Byzantium. But the event itself, the excommunication of the patriarch Michael Cerularius by Cardinal Humbert in Constantinople, symbolized an irreconcilable difference in ideology. The reform movement in the Roman Church had emphasized an ideal of the universal role of the papacy that was wholly incompatible with Byzantine tradition. Both sides also deliberately aggravated their differences by reviving all the disputed points of theology and ritual that had become battle cries during the Photian Schism in the 9th century. The schism of 1054 passed unnoticed by contemporary Byzantine historians; its significance as a turning point in East–West relations was fully realized only later.

Schism between churches

Alexius I and the First Crusade. But even the events of 1071 had not made the decline of Byzantium irretrievable. The shrinking of its boundaries reduced the empire from its status as a dominating world power to that of a small Greek state fighting for survival. That survival now depended on the new political, commercial, and ecclesiastical forces in the West, for it could no longer draw on its former military and economic resources in Anatolia. The civil aristocracy of Constantinople yielded with bad grace. After four years of civil war, the military lords triumphed with the accession of Alexius I Comnenus, the greatest soldier and statesman to hold the throne since Basil II. The history of his reign was written in elegant Greek by his daughter Anna Comnena; and, as she remarks, it began with an empire beset by enemies on all sides. The Normans captured Dyrrhachium (modern Durrës, Alb.) in 1082 and planned to advance overland to Thessalonica. Alexius called on the Venetians to help him, but Robert Guiscard's death in 1085 temporarily eased the Norman problem. The following year the Seljuq sultan died, and the sultanate was split by internal rivalries. Fortune thus played into Alexius's hands by ridding him of two of his besetting enemies. By his own efforts, however, he defeated the Pechenegs in 1091.

The Venetians had been pleased to help drive the Normans out of the Adriatic Sea but demanded a heavy price. In 1082 Alexius granted them trading privileges in Constantinople and elsewhere on terms calculated to outbid Byzantine merchants. This charter was the cornerstone of the commercial empire of Venice in the eastern Mediterranean. But it fed the flames of Byzantine resentment against the Latins; and it provoked the rich, who might have been encouraged to invest their capital in shipbuilding and trade, to rely on the more familiar security of landed property.

Involvement in the First Crusade

The terms that Alexius made with his enemies in the first 10 years of his reign were not meant to be permanent. He fully expected to win back Anatolia from the Seljuqs; his plans, however, were not given time to mature, for matters were precipitated by the arrival in the East of the first crusaders from western Europe (1096). Alexius had undoubtedly solicited the help of mercenary troops from the West but not for the liberation of the Holy Land from the infidel. The urgent need was the protection of Constantinople and the recovery of Anatolia. The Byzantines were more realistic about their Muslim neighbours than the distant popes and princes of the West. Jerusalem had finally been taken by the Seljuqs in 1071, but the most immediate threat to Byzantium came from the Pechenegs and the Normans. Alexius was tactful in his dealings with the pope and ready to discuss the differences between the churches. But neither party foresaw the consequences of Pope Urban II's appeal in 1095 for recruits to fight a Holy War. The response in western Europe was overwhelming. The motives of those who took the cross as crusaders ranged from religious enthusiasm to a mere spirit of adventure or a hope of gain; and it was no comfort to Alexius to learn that four of the eight leaders of the First Crusade were Normans, among them Bohemond, the son of Robert Guiscard. Since the crusade had to pass through Constantinople, however, the Emperor had some control over it. He required its leaders to swear to restore to the empire any towns or territories they might conquer from the Turks on their way to the Holy Land. In return, he gave them guides and a military escort. Still, the cost was enormous, for the crusaders had to be supplied with food or live off the land as they went.

Nicaea fell to them in 1097 and was duly handed over to the Emperor in accord with the agreement. In 1098 they reached, and captured, Antioch. There the trouble started. Bohemond refused to turn over the city and instead set up his own principality of Antioch. His example was imitated in the establishment of the Latin Kingdom of Jerusalem (1100), which had fallen to the crusaders the year before, and of the counties of Edessa and Tripoli. The crusaders settled down to colonize and defend the coast of Palestine and Syria and to quarrel among themselves. While they did so, Alexius was able to establish a new and more secure boundary between Byzantium and Islām through the middle of Anatolia. Full advantage was taken of the prevailing rivalry between the Seljuq sultans at Konya and the rival dynasty of the Dānishmend emirs at Melitene (near modern Malatya, Tur.); and a limit was set to the westward expansion of the Turks.

The *pronoia* system

The First Crusade thus brought some benefits to Byzantium. But nothing could reconcile the emperor to Bohemond of Antioch. In 1107 Bohemond mounted a new invasion of the empire from Italy. Alexius was ready and defeated him at Dyrrhachium in 1108. Byzantine prestige was higher than it had been for many years, but the empire could barely afford to sustain the part of a great power. Alexius reconstituted the army and re-created the fleet, but only by means of stabilizing the gold coinage at one-third of its original value and by imposing a number of supplementary taxes. It became normal practice for taxes to be farmed out, which meant that the collectors recouped their outlay on their own terms. People in the provinces had the added burden of providing materials and labour for defense, communications, and provisions for the army, which now included very large numbers of foreigners. The supply of native soldiers had virtually ceased with the disappearance or absorption of their military holdings. Alexius promoted an alternative source of native manpower by extending the system of granting estates in *pronoia* (by favour of the emperor) and tying the grant to the military obligation. The recipient of a *pronoia* was entitled to all the revenues of his estate and to the taxes payable by his tenants (*paroikoi*), on condition of equipping himself as a mounted cavalryman with a varying number of troops. He was in absolute possession of his property until it reverted to the crown upon his death. Similarly, Alexius tried to promote more profitable development of the estates of the church by granting them to the management of laymen as *charistikia* or benefices. As an expedient, the *pronoia* system had advantages both for the state and for the military aristocracy who were its main beneficiaries. But in the long term it hastened the fragmentation of the empire among the landed families and the breakdown of centralized government that the 10th-century emperors had laboured to avert.

Later Comneni. The policies of Alexius I were continued by his son John II Comnenus (reigned 1118–43) and his grandson Manuel I Comnenus (reigned 1143–80). In the 12th century there was a growing involvement of the Western powers in the affairs of the East, as well as an increasingly complex political situation in Europe. In Asia, too, matters were complicated by the conflict between the Seljuqs and the Dānishmends, by the emergence of the kingdom of Lesser Armenia in Cilicia, and by the activities of the crusader states. Foreign relations and skillful diplomacy became of paramount importance for the Byzantines. John II tried and failed to break what was becoming the Venetian monopoly of Byzantine trade, and he sought to come to terms with the new kingdom of Hungary, to whose ruler he was related by marriage. Alexius I had seen the importance of Hungary, lying between the Western and Byzantine empires, a neighbour of the Venetians and the Serbs. More ominous still was the establishment of the Norman Kingdom of Sicily under Roger II in 1130. But John II astutely allied himself with the Western emperor against it.

Appeasement of the West

Manuel I realized even more clearly that Byzantium could not presume to ignore or offend the new powers in the West, and he went out of his way to understand and to appease them. Certain aspects of the Western way of life appealed to Manuel. His first and second wives were both Westerners, and Latins were welcomed at his court and even granted estates and official appointments. This policy was distasteful to most of his subjects; and it was unfortunate for his intentions that the Second Crusade occurred early in his reign (1147), for it aggravated the bitterness between Greeks and Latins and brought Byzantium deeper than ever into the tangled politics of western Europe. Its leaders were Louis VII of France and the emperor Conrad III, and its failure was blamed on Byzantine treachery. The French king discussed with Roger of Sicily the prospect of attacking Constantinople, and in 1147 Roger invaded Greece. But Manuel retained the personal friendship and the alliance of Conrad III against the Normans and even planned a joint Byzantine-German campaign against them in Italy.

No such cooperation was possible with Conrad's successor, Frederick I Barbarossa, after 1152. To Frederick, the alliance between the Holy Roman Empire and what he called "the kingdom of the Greeks" was not one between equals. Manuel launched a vain invasion of the Norman kingdom on his own account in 1154, but it was too late for a revival of Byzantine imperialism in the West. It was hard for the Byzantines to accept the fact that their empire might soon become simply one among a number of Christian principalities.

In the Balkans and in the Latin East Manuel was more successful. His armies won back much of the northwest Balkans and almost conquered Hungary, reducing it to a client kingdom of Byzantium. The Serbs, too, under their leader Stephen Nemanja, were kept under control, while Manuel's dramatic recovery of Antioch in 1159 caused the crusaders to treat the Emperor with a new respect. But in Anatolia he overreached himself. To forestall the formation of a single Turkish sultanate, Manuel invaded the Seljuq territory of Rūm in 1176. His army was surrounded and annihilated at Myriocephalon. The battle marked the end of the Byzantine counteroffensive against the Turks begun by Alexius I. Its outcome delighted the Western emperor, Frederick I Barbarossa, who had supported the Seljuq sultan of Rūm against Manuel and who now openly threatened to take over the Byzantine Empire by force.

Manuel's personal relationships with the crusaders and with other Westerners remained cordial to the end. But his policies had antagonized the Holy Roman Empire, the papacy, the Normans, and, not least, the Venetians. His effort to revive Byzantine prestige in Italy and the Balkans, and his treaties with Genoa (1169) and Pisa (1170), roused the suspicions of Venice; and in 1171, following an anti-Latin demonstration in Constantinople, all Venetians in the empire were arrested and their property was confiscated. The Venetians did not forget this episode. They, too, began to think in terms of putting Constantinople under Western control as the only means of securing their interest in Byzantine trade.

Manuel's policies antagonized many of his own people as well. His favouritism to the Latins was unpopular, as was his lavish granting of estates in *pronoia.* A reaction set in shortly after his death in 1180, originated by his cousin Andronicus I Comnenus, who ascended to the throne after another anti-Latin riot in Constantinople. Andronicus murdered Manuel's widow and son Alexius II. He posed as the champion of Byzantine patriotism and of the oppressed peasantry. But to enforce his reforms he behaved like a tyrant. By undermining the power of the aristocracy he weakened the empire's defenses and undid much of Manuel's work. The King of Hungary broke his treaty, and Stephen Nemanja of Serbia declared his independence from Byzantium and founded a new Serbian kingdom. Within the empire, too, disintegration proceeded. In 1185 Isaac Comnenus, governor of Cyprus, set himself up as independent ruler of the island. In the same year the Normans again invaded Greece and captured Thessalonica. The news prompted a counterrevolution in Constantinople, and Andronicus was murdered.

He was the last of the Comnenian family to wear the crown. His successor, Isaac II Angelus, was brought to power by the aristocracy. His reign, and, still more, that of his brother Alexius III, saw the collapse of what remained of the centralized machinery of Byzantine government and defense. Isaac tried at least to keep his foreign enemies in check. The Normans were driven out of Greece in 1185. But in 1186 the Bulgars began a rebellion that was to lead to the formation of the Second Bulgarian Empire. Matters were not made easier by the arrival of the Third Crusade, provoked by the loss of Jerusalem to the Muslim leader Saladin in 1187. One of its leaders was Frederick I Barbarossa, whose avowed intention was to conquer Constantinople. He died on his way to Syria. But Richard I the Lion-Heart of England appropriated Cyprus from Isaac Comnenus, and the island never again reverted to Byzantine rule.

The Fourth Crusade and the establishment of the Latin Empire. In 1195 Isaac II was deposed and blinded by his brother Alexius III. The Westerners, who had again blamed the failure of their crusade on the Byzantines, saw ways of exploiting the situation. The emperor Henry VI had united the Norman Kingdom of Sicily with the Holy Roman Empire. He inherited the ambitions of both to master Constantinople, and his brother, Philip of Swabia, was married to a daughter of the dethroned Isaac II. Alexius bought off the danger by paying tribute to Henry, but Henry died in 1197. The idea had now gained ground in the West that the conquest of Constantinople would solve a number of problems and would be of benefit not only to trade but also to the future of the crusade and the church. In 1198 Innocent III was elected pope. The new rulers of Hungary, Serbia, and Bulgaria all turned to him for the recognition of the sovereignty that Byzantium would not give them.

Conquest and colonization by the West

It was under Innocent's inspiration that the Fourth Crusade was launched, and it was by the diversion of that crusade from its purpose and objective that the conquest and colonization of the Byzantine Empire by the West was realized. A multiplicity of causes and coincidences led up to the event, but the ambition of Venice, which supplied the ships, must rank high among them. A plausible excuse was offered by the cause of restoring Isaac II, whose son Alexius IV had escaped to the West to seek help, and who made lavish promises of reward to his benefactors. But when, in 1203, the crusaders drove Alexius III out of Constantinople, Isaac II and his son proved incapable either of fulfilling the promises or of stifling the anti-Latin prejudice of their people, who proclaimed an emperor of their own in the person of Alexius V. The Venetians and crusaders therefore felt justified in taking their own reward by conquering and dividing Constantinople and the Byzantine provinces among themselves. The city fell to them in April 1204. They worked off their resentment against the inhabitants in an unparalleled orgy of looting and destruction, which did irreparable damage to the city and immeasurable harm to East–West understanding.

The Venetians, led by their doge, Enrico Dandolo, gained most from the enterprise by appropriating the principal harbours and islands on the trade routes. The crusaders set about the conquest of the European and Asiatic provinces. The first Latin emperor, Baldwin I, was the suzerain of the feudal principalities that they established in Thrace, Thessalonica, Athens, and the Morea (Peloponnese). He soon came into conflict with the ruler of Bulgaria. Still more serious was the opposition offered by the three provincial centres of Byzantine resistance. At Trebizond (Trabzon) on the Black Sea, two brothers of the Comnenian family laid claim to the imperial title. In Epirus in northwestern Greece Michael Angelus Ducas, a relative of Alexius III, made his capital at Arta and harassed the crusader states in Thessaly. The third centre of resistance was based on the city of Nicaea in Anatolia, where Theodore I Lascaris, another relative of Alexius III, was crowned as emperor in 1208 by a patriarch of his own making. Of the three, Nicaea lay nearest to Constantinople, between the Latin Empire and the Seljuq sultanate of Rūm; and its emperors proved worthy of the Byzantine traditions of fighting on two fronts at once and of skillful diplomacy. Theodore Lascaris and his son-in-law John III Vatatzes built up at Nicaea a microcosm of the Byzantine Empire and church in exile. The Latins were thus never able to gain a permanent foothold in Anatolia; and even in Europe their position was constantly threatened by the Byzantine rulers of northern Greece, though in the centre and south of the country their conquests were more lasting.

The most successful of the Latin emperors was Baldwin's brother, Henry of Flanders, after whose death in 1216 the Latin Empire lost the initiative and the recovery of Constantinople became a foreseeable goal for the Byzantines in exile. The Latin regime was prolonged less by its own vitality than by the inability of the successor states of Epirus and Nicaea to cooperate. In 1224 Theodore Ducas of Epirus, who had extended his territories across the north of Greece and far into Bulgaria, wrested Thessalonica from the Latins and was crowned emperor there in defiance of the Emperor in Nicaea. In 1230, however, he was defeated in battle against the Bulgars before reaching Constantinople; and his defeat gave John III Ducas Vatatzes the chance to extend his own empire into Europe, to ally with the Bulgars, and so to encircle Constantinople. Theodore's successor was made to renounce his imperial title, and Thessalonica surrendered to the empire of Nicaea in 1246. The Mongol invasion of Anatolia, which had meanwhile thrown the East into confusion, was of great benefit to Nicaea, for it weakened the Seljuq sultanate and isolated the rival empire of Trebizond.

The accession of Michael Palaeologus

John Vatatzes might well have crowned his achievements by taking Constantinople had he not died in 1254. When his son Theodore II Lascaris (1254–58) died in 1258, leaving an infant son, John IV, the regency and then the throne in Nicaea were taken over by Michael VIII Palaeologus (reigned 1259–82). Michael came from one of the aristocratic families of Nicaea whom Theodore II had mistrusted. But it was he who carried the work of the Lascarid emperors to its logical conclusion. The Byzantine state in Epirus had revived under Michael II Ducas, who set his sights on Thessalonica. Despite several efforts to reach a diplomatic settlement, the issue between the rival contenders had finally to be resolved in battle at Pelagonia in Macedonia in 1259. Michael II was supported by William of Villehardouin, the French prince of the Morea, and by Manfred, the Hohenstaufen king of Sicily. The victory went to the army of Nicaea. Two years later a general of that army entered Constantinople. The last of the Latin emperors, Baldwin II, fled to Italy; and the Venetians were dispossessed of their lucrative commercial centre. In August 1261 Michael VIII was crowned as emperor in Constantinople; the boy heir to the throne of Nicaea, John IV Lascaris, was blinded and imprisoned. In this way, the dynasty of Palaeologus, the last to reign in Constantinople, was inaugurated.

THE EMPIRE UNDER THE PALAEOLOGI: 1261–1453

The empire in exile at Nicaea had become a manageable and almost self-sufficient unit, with a thriving economy based on agriculture and, latterly, on trade with the Seljuqs. It had no navy but the land frontiers in Anatolia, policed by well-paid troops, were stronger than they had been since the 12th century. By stretching the frontiers into Europe the empire had not dissipated its strength; for the possession of Thessalonica balanced that of Nicaea. When the seat of government was moved from Nicaea to Constantinople, that balance was upset, the economy was re-oriented, and the defense system in Anatolia began to break down. Constantinople was still the New Jerusalem for the Byzantines. To leave it in foreign hands was unthinkable. But after the dismemberment of the empire by the Fourth Crusade, the city was no longer the focal point of an integrated structure. It was more like an immense city-state in the midst of a number of more or less independent provinces. Much of Greece and the islands remained in French or Italian hands. The Byzantine rulers of Epirus and Thessaly, like the emperors in Trebizond, refused to recognize Michael VIII as emperor. His treatment of the Lascarid heir of Nicaea, for which the patriarch Arsenius excommunicated him, appalled many of his own subjects and provoked what was known as the Arsenite schism in the Byzantine Church. Many in Anatolia, loyal to the memory of the Lascarid emperors who had enriched and protected them, condemned Michael VIII as a usurper.

Michael VIII. The new dynasty was thus founded in an atmosphere of dissension, but its founder was determined that it should succeed. He took measures for the rehabilitation, repopulation, and defense of Constantinople. He stimulated a revival of trade by granting privileges to Italian merchants. The Genoese, who had agreed to lend him ships for the recovery of the city from their Venetian rivals, were especially favoured; and soon they had built their own commercial colony at Galata opposite Constantinople, and cornered most of what had long been a Venetian monopoly. Inevitably, this led to a conflict between Genoa and Venice, of which the Byzantines were the main victims. Some territory was taken back from the Latins, notably in the Morea and the Greek islands. But little was added to the imperial revenue; and Michael VIII's campaigns there and against Epirus and Thessaly ate up the resources that had been accumulated by the emperors at Nicaea.

Threat of reconquest by the West

The dominating influence on Byzantine policy for most of Michael's reign was the threat of reconquest by the Western powers. Charles of Anjou, the brother of the French king Louis IX, displaced Manfred of Sicily and inherited his title in 1266; he then organized a coalition of all parties interested in re-establishing the Latin empire, posing as the pope's champion to lead a crusade against the schismatic Greeks. Michael VIII countered this threat by offering to submit the Church of Constantinople to the see of Rome, thereby inviting the pope's protection

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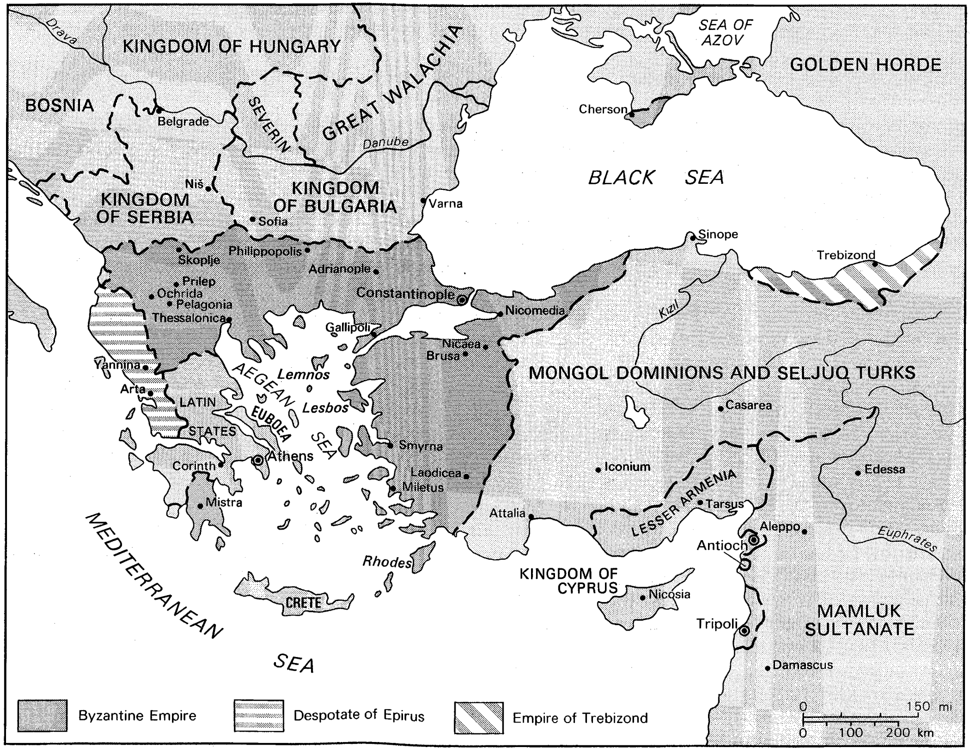

The remnants of the Byzantine Empire in 1265.

and removing the only moral pretext for a repetition of the Fourth Crusade. The offer to reunite the churches had been made as a diplomatic ploy to previous popes by previous emperors, but never in such compelling circumstances. Pope Gregory X accepted it at its face value, and at the second Council of Lyon in 1274 a Byzantine delegation professed obedience to the Holy See in the name of their emperor. Michael's policy, sincere or not, was violently opposed by most of his people, and he had to persecute and imprison large numbers of them in order to persuade the papacy that the union of the churches was being implemented. Later popes were not convinced by the pretense. In 1281 Charles I (Charles of Anjou) invaded the empire. His army was beaten back in Albania, but he at once prepared a new invasion by sea, supported by Venice, Serbia, Bulgaria, and the separatist rulers of northern Greece. His plans, however, were wrecked in 1282 by a rebellion in Sicily called the Sicilian Vespers and by the intervention of Peter III of Aragon, which the Byzantines encouraged. Michael VIII died at the end of the same year. He had saved his empire from its most persistent enemy, but he died condemned by his church and people as a heretic and a traitor.

Neglect of eastern defenses

Whatever sins he may have committed in the eyes of the Orthodox Church, it is true that Michael VIII, by concentrating on the danger from the West, neglected, if he did not betray, the eastern provinces where he had come to power. Frontier defense troops in Anatolia were withdrawn to Europe or neglected, and bands of Turkish raiders, driven westward by the upheaval of the Mongol invasion, began to penetrate into Byzantine territory. Like the Seljuqs in the 11th century, the new arrivals found little organized opposition. Some of the local Byzantines even collaborated with them out of their own antipathy to the Emperor in Constantinople. By about 1280 the Turks were plundering the fertile valleys of western Anatolia, cutting communications between the Greek cities, and their emirs were beginning to carve out small principalities. Michael VIII's network of diplomacy covered the Mongols of Iran and the Golden Horde in Russia, as well as the Mamlūks of Egypt. But diplomacy was ineffective against Muslim Ghazis (warriors inspired by the ideal of holy war); by the time the threat from Italy was removed in 1282, it was almost too late to save Byzantine Anatolia.

Nor was it possible to raise armies to fight in Europe and Asia simultaneously. The native recruitment fostered by the Comnenian emperors had fallen off since 1261. Estates held in *pronoia* had become hereditary possessions of their landlords, who ignored or were relieved of the obligation to render military service to the government. The knights of the Fourth Crusade had found many familiar elements of feudalism in the social structure of the Byzantine provinces. By the end of the 13th century the development had gone much further. The officers of the Byzantine army were still mostly drawn from the native aristocracy. But the troops were hired, and the cost of maintaining a large army in Europe, added to the lavish subsidies that Michael VIII paid to his friends and allies, crippled the economy.

Andronicus II. Michael's son Andronicus II (reigned 1282–1328) unwisely attempted to economize by cutting down the size of the army and disbanding the navy. Unemployed Byzantine sailors sold their services to the new Turkish emirs, who were already raiding the Aegean islands. The Genoese became the suppliers and defenders of Constantinople by sea, which excited the jealousy of the Venetians to the pitch of war and led to the first of a series of naval battles off Constantinople in 1296. In reaction against his father's policy, Andronicus II pursued a line of almost total isolation from the papacy and the West. The union of Lyon was solemnly repudiated and Orthodoxy restored, to the deep satisfaction of most Byzantines. But there were still divisive conflicts in society. The Arsenite schism in the church was not healed until 1310; the rulers of Epirus and Thessaly remained defiant and kept contact with the successors of Charles I in Italy; and the people of Anatolia aired their grievances in rebellion. As the Turks encroached on their land, refugees in growing numbers fled to the coast or to Constantinople, bringing new problems for the government. In 1302 a band of Turkish warriors defeated the Byzantine army near Nicomedia in northwestern Anatolia. Its leader, Osman I, was the founder of the Osmanli, or Ottoman, people, who were soon to overrun the Byzantine Empire in Europe.

Appearance of Ottoman Turks

In 1303 Andronicus hired a professional army of mercenaries, the Grand Catalan Company. The Catalans made one successful counterattack against the Turks in Anatolia. But they were unruly and unpopular, and when their leader was murdered they turned against their employers. For some years they used the Gallipoli Peninsula as a base from which to ravage Thrace, inviting thousands of Turks to come over and help them. The Catalans finally moved west; in 1311 they conquered Athens from the French and established the Catalan Duchy of Athens and Thebes. The Turks whom they left behind were not ejected from Gallipoli until 1312. The cost of hiring the Catalans, and then of repairing the damage that they had done, had to be met by desperate measures. The face value of the Byzantine gold coin, the *hyperpyron,* was lowered when its gold content was reduced to a mere 50 percent; and the people had to bear still greater burdens of taxation—some payable in kind by farmers. Inflation and rising prices led to near famine in Constantinople, the population of which was swollen by vast numbers of refugees.

Cultural revival. Materially, the empire seemed almost beyond hope of recovery in the early 14th century, but spiritually and culturally it showed a remarkable vitality. The church, no longer troubled over the question of union with Rome, grew in prestige and authority. The patriarchs of Constantinople commanded the respect of all the Orthodox churches, even beyond the imperial boundaries; and Andronicus II, himself a pious theologian, yielded to the patriarch the ancient right of imperial jurisdiction over the monastic settlement on Mt. Athos. There was a new flowering of the Byzantine mystical tradition in a movement known as Hesychasm, whose chief spokesman was Gregory Palamas, a monk from Athos. The theology of the Hesychasts was thought to be heterodox by some theologians, and a controversy arose in the second quarter of the 14th century that had political undertones and was as disruptive to the church and state as the Iconoclastic dispute had been in an earlier age. It was not resolved until 1351.

Mysticism and scholarship

The revival of mystical speculation and the monastic life may have been in part a reaction against the contemporary revival of secular literature and learning. Scholarship of all kinds was patronized by Andronicus II. As in the 11th century, interest was mainly centred on a rediscovery of ancient Greek learning. The scholar Maximus Planudes compiled a famous anthology and translated a number of Latin works into Greek, though knowledge of Latin was rare and most of the Byzantine scholars prided themselves on having in their Hellenic heritage an exclusive possession that set them apart from the Latins. A notable exception was Demetrius Cydones who, like Michael Psellus, managed affairs of state for a number of emperors for close to 50 years. Cydones translated the works of Thomas Aquinas into Greek; he was the forerunner of a minority of Byzantine intellectuals who joined the Roman Church and looked to the West to save their empire from ruin. More typical of his class was Theodore Metochites, the Grand Logothete, or chancellor, of Andronicus II, whose encyclopaedic learning rivaled that of Psellus. His pupil Nicephorus Gregoras, in addition to his researches in philosophy, theology, mathematics, and astronomy, wrote a history of his age. The tradition of Byzantine historiography, maintained by George Acropolites, the historian of the Empire of Nicaea, was continued in the 14th century by George Pachymeres, by Gregoras, and finally by the emperor John VI Cantacuzenus, who wrote his memoirs after his abdication in 1354.

Andronicus III and John Cantacuzenus. The histories they wrote tell more of politics and personalities than of the underlying social and economic tensions in their society that were to find expression in a series of civil wars. Trouble broke out in 1320 when Andronicus II, purely for family reasons, disinherited his grandson Andronicus III. The cause of the young emperor was taken up by his friends, and there was periodic warfare from 1321 to 1328, when the older Andronicus had to yield the throne. It was in some ways a victory for the younger generation of the aristocracy, of whom the leading light was John Cantacuzenus. It was he who guided the empire's policies during the reign of Andronicus III (1328–41). They were men of greater drive and determination, but the years of fighting had made recovery still more difficult and had given new chances to their enemies. In 1329 they fought and lost a battle at Pelekanon (near Nicomedia) against Osman's son, Orhan, whose Turkish warriors went on to capture Nicaea in 1331 and Nicomedia in 1337. Northwestern Anatolia, once the heart of the empire, was now lost. There seemed no alternative but to accept the fact and to come to terms with the Ottomans and the other Turkish emirs. By so doing, Andronicus III and Cantacuzenus were able to call on the services of almost limitless numbers of Turkish soldiers to fight for them against their other enemies: the Italians in the Aegean islands and the Serbs and the Bulgars in Macedonia and Thrace.

The power of Serbia, which Andronicus II had managed to control by diplomatic means, grew alarmingly after the accession of Stefan Dušan to the Serbian throne in 1331. Dušan exploited to the full the numerous embarrassments of the Byzantines and in 1346 announced his ambitions by having himself crowned as emperor of the Serbs and Greeks. The greatest practical achievement of Andronicus III was the restoration to Byzantine rule of the long-separated provinces of Epirus and Thessaly. But only a few years later, in 1348, the whole of northern Greece was swallowed up in the Serbian Empire of Stefan Dušan.

Outbreak of civil war

When Andronicus III died in 1341, civil war broke out for a second time. The contestants on that occasion were John Cantacuzenus, who had expected to act as regent for the boy-heir John V, and his political rivals led by his former partisan Alexius Apocaucus, the patriarch John Calecas, and the empress mother Anne of Savoy, who held power in Constantinople. Cantacuzenus, befriended and then rejected by Dušan of Serbia, was crowned as Emperor John VI in Thrace in 1346; and, with the help of Turkish troops, he fought his way to victory in the following year. Like Romanus Lecapenus, he protested that he was no more than the protector of the legitimate heir to the throne, John V Palaeologus. His brief reign, from 1347 to 1354, might have turned the tide of Byzantine misfortunes had not the second civil war provoked unprecedented social and political consequences. In the cities of Thrace and Macedonia the people vented their dissatisfaction with the ruling aristocracy by revolution. It was directed mainly against Cantacuzenus and the class that he represented. The movement was most memorable and lasting in Thessalonica, where a faction known as the Zealots seized power in a coup d'état and governed the city as an almost independent commune until 1350.

The second civil war was consequently even more destructive of property and ruinous to the economy than the first. At the same time, in 1347, the Black Death decimated the population of Constantinople and other parts of the empire. John VI Cantacuzenus, nevertheless, did what he could to restore the economy and stability of the empire. To coordinate the scattered fragments of its territory he assigned them as appanages to individual members of the imperial family. His son Manuel took over the province of the Morea in 1349 with the rank of despot and governed it with growing success until his death in 1380; his eldest son, Matthew, was given a principality in Thrace; while the junior emperor John V, who had married a daughter of Cantacuzenus, ruled in Thessalonica after 1351.

Cantacuzenus also tried but failed to weaken the economic stranglehold of the Genoese by rebuilding a Byzantine war fleet and merchant navy. The effort involved him in warfare, first on his own and then as an unwilling partner of the Venetians against the Genoese, from which Byzantium emerged as the loser. The revenue of the Genoese colony at Galata, derived from custom dues, was now far greater than that of Constantinople. The empire's poverty was reflected in dilapidated buildings and falling standards of luxury. The crown jewels had been pawned to Venice during the civil war, and the Byzantine gold coin, hopelessly devalued, had given place in international trade to the Venetian ducat. More and more, Byzantium was at the mercy of its foreign competitors and enemies, who promoted and exploited the political and family rivalries among the ruling class. John Cantacuzenus was never popular as an emperor, and feeling against him came to a head when some of his Ottoman mercenaries took the occasion of the destruction of Gallipoli by earthquake to occupy and fortify the city in March 1354. It was their first permanent establishment in Europe, at the key point of the crossing from Asia. In November of the same year John V

The Byzantine Empire in 1355.

Turkish occupation of Gallipoli

Palaeologus, encouraged by the Anti-Cantacuzenist Party, forced his way into Constantinople. In December Cantacuzenus abdicated and became a monk. Though his son Matthew, who had by then been crowned as coemperor, fought on for a few years, the dynasty of Cantacuzenus was not perpetuated.

Turkish expansion. John Cantacuzenus' relationship with the Turks had been based on personal friendship with their leaders, among them Orhan, to whom he gave his daughter in marriage. But once the Turks had set up a base on European soil and had seen the possibilities of further conquest, such relationships were no longer practicable. Stefan Dušan, who very nearly realized his ambition to found a new Serbo-Byzantine empire, was the only man who might have prevented the subsequent rapid expansion of the Turks into the Balkans, but he died in 1355 and his empire split up. The new emperor, John V, hoped that the Western world would sense the danger, and in 1355 he addressed an appeal for help to the Pope. The popes were concerned for the fate of the Christian East but guarded in their offers to Constantinople so long as the Byzantine Church remained in schism from Rome. In 1366 John V visited Hungary to beg for help, but in vain. In the same year his cousin Amadeo, count of Savoy, brought a small force to Constantinople and recaptured Gallipoli from the Turks, who had by then advanced far into Thrace. Amadeo persuaded the Emperor to go to Rome and make his personal submission to the Holy See in 1369. On his way home, John was detained at Venice as an insolvent debtor; during his absence the Turks scored their first victory over the successors of Stefan Dušan on the Marica River near Adrianople in 1371. The whole of Macedonia was open to them. The remaining Serbian princes and the ruler of Bulgaria became their vassals, and in 1373 the Emperor was forced to do the same.

Byzantium became a vassal state of the Turks, pledged to pay tribute and to provide military assistance to the Ottoman sultan. The possession of Constantinople thereafter was disputed by the Emperor's sons and grandsons in a series of revolutions, which were encouraged and sometimes instigated by the Turks, the Genoese, or the Venetians. John V's son Andronicus IV, aided by the Genoese and the sultan Murad I, mastered the city for three years (1376–79). He rewarded the Turks by giving back Gallipoli to them, and Murad made his first European capital at Adrianople. The Venetians helped John V to regain his throne in 1379, and the empire was once again divided into appanages under his sons. Only his second son, Manuel, showed any independence of action. For nearly five years, from 1382 to 1387, Manuel reigned as emperor at Thessalonica and laboured to make it a rallying point for resistance against the encroaching Turks. But the city fell to Murad's army in April 1387. When the Turks then drove deeper into Macedonia, the Serbs again organized a counteroffensive but were overwhelmed at Kossovo in 1389.

Manuel II and respite from the Turks. The loss of Thessalonica and the Battle of Kossovo sealed off Constantinople by land. The new sultan Bayezid I (1389–1402) intended to make it his capital; when Manuel II came to that throne at his father's death in 1391, the Sultan warned him that he was emperor only inside the city walls. The Turks already controlled the rest of Byzantine Europe, except for the south of Greece.

In 1393 Bayezid completed his conquest of Bulgaria, and soon afterward he laid siege to Constantinople. The blockade was to last for many years. Manuel II, like his father, pinned his hopes of rescue on the West. A great crusade against the Turks was organized by the King of Hungary, but it was defeated at Nicopolis on the Danube in 1396. In 1399 the French marshal Boucicaut, who had been at Nicopolis and had returned to the relief of Constantinople with a small army, persuaded Manuel to travel to western Europe to put the Byzantine case in person. From the end of 1399 to June 1403 the Emperor visited in Italy, France, and England, leaving his nephew John VII in charge of Constantinople. Manuel's journey did something to stimulate Western interest in Greek learning. His friend and ambassador in the West, Manuel Chrysoloras, a pupil of Demetrius Cydones, was appointed to teach Greek at Florence. The Pope instituted a defense fund for Constantinople. Interest and sympathy were forthcoming but little in the way of practical help. During Manuel's absence, however, the Ottomans were defeated at Ankara by the Mongol leader Timur (Tamerlane) in July 1402. Bayezid was captured and his empire in Asia was shattered. His four sons contended with each other to secure possession of the European provinces, which had been little affected by the Mongol invasion, and to reunite the Ottoman dominions. In these wholly unexpected circumstances the Byzantines found themselves the favoured allies first of one Turkish contender, then of another. The blockade of Constantinople was lifted. Thessalonica—with Mt. Athos and other places—was restored to Byzantine rule, and the payment of tribute to the sultan was annulled. In 1413 Mehmed I, helped and promoted by the emperor Manuel, triumphed over his rivals and became sultan of the reintegrated Ottoman Empire.

Appeals of Manuel II in western Europe

During his reign, from 1413 to 1421, the Byzantines enjoyed their last respite. Manuel II, aware that it could not last, made the most of it by strengthening the defenses and administration of the fragments of his empire. The most flourishing province in the last years was the Despotate of Morea. Its prosperity had been built up first by the sons of John Cantacuzenus (who died there in 1383) and then by the son and grandson of John V—Theodore I and Theodore II Palaeologus. Its capital city of Mistra became a haven for Byzantine scholars and artists and a centre of the last revival of Byzantine culture, packed with churches, monasteries, and palaces. Among its scholars was George Gemistus Plethon, a Platonist who dreamed of a rebirth of Hellenism on Hellenic soil.

Final Turkish assault. When Murad II became sultan, in 1421, the days of Constantinople and of Hellenism were numbered. In 1422 Murad revoked all the privileges accorded to the Byzantines by his father and laid siege to Constantinople. His armies invaded Greece and blockaded Thessalonica. The city was then a possession of Manuel II's son Andronicus, who in 1423 handed it over to the Venetians. For seven years Thessalonica was a Venetian colony, until, in March 1430, the Sultan assaulted and captured it. Meanwhile, Manuel II had died in 1425, leaving his son John VIII as emperor. John, who had already traveled to Venice and Hungary in search of help, was prepared to reopen negotiations for the union of the churches as a means of stirring the conscience of Western Christendom. His father had been skeptical about the benefits of such a policy, knowing that it would antagonize most of his own people and arouse the suspicion of the Turks. The proposal was made, however, at the Council of Florence in 1439, attended by the emperor John VIII, his patriarch, and many Orthodox bishops and dignitaries. After protracted and difficult discussions, they agreed to submit to the authority of Rome. The union of Florence was badly received by the citizens of Constantinople and by most of the Orthodox world. But it had its notable adherents, such as the bishops Bessarion of Nicaea and Isidore of Kiev, both of whom retired to Italy as cardinals of the Roman Church. Bessarion's learning and library helped to encourage further Western interest in Greek scholarship. The union of Florence also helped to stimulate a crusade against the Turks. Once again it was led by the king of Hungary, Władysław III of Poland, supported by George Branković of Serbia and by János Hunyadi of Transylvania. But there were disagreements among its leaders, and the Christian army was annihilated at Varna in 1444.

Fall of Thessalonica

The Byzantine collapse and the Ottoman triumph followed swiftly thereafter. In 1448 Constantine XI (or XII), the last emperor, left Mistra for Constantinople when his brother John VIII died without issue. His two other brothers, Thomas and Demetrius, continued to govern the Morea, the last surviving Byzantine province. In 1449 Mehmed II (sultan 1444–46 and 1451–81) began to prepare for the final assault on Constantinople. No further substantial help came from the West, and the formal celebration of the union of the churches in Hagia Sophia in 1452 was greeted with a storm of protest. Even in their extremity, the Byzantines would not buy their freedom at the expense of their Orthodox faith. They found the prospect of being ruled by the Turks less odious than that of being indebted to the Latins. When the crisis came, however, the Venetians in Constantinople, and a Genoese contingent commanded by Giovanni Giustiniani, wholeheartedly cooperated in the defense of the city. Mehmed II laid siege to the walls in April 1453. His ships were obstructed by a chain that the Byzantines had thrown across the mouth of the Golden Horn. The ships were therefore dragged overland to the harbour from the seaward side, bypassing the defenses. The Sultan's heavy artillery continually bombarded the land walls until, on May 29, some of his soldiers forced their way in. Giustiniani was mortally wounded. The emperor Constantine was last seen fighting on foot at one of the gates.

The Sultan allowed his victorious troops three days and nights of plunder before he took possession of his new capital. The Ottoman Empire had now superseded the Byzantine Empire; and some Greeks, like the contemporary historian Critobulus of Imbros, recognized the logic of the change by bestowing on the Sultan all the attributes of the emperor. The material structure of the empire, which had long been crumbling, was now under the management of the sultan-basileus. But the Orthodox faith was less susceptible to change. The Sultan acknowledged the fact that the church had proved to be the most enduring element in the Byzantine world, and he gave the Patriarch of Constantinople an unprecedented measure of temporal authority by making him answerable for all Christians living under Ottoman rule.

The last scattered pockets of Byzantine resistance were eliminated within a decade after 1453. Athens fell to the Turks in 1456–58, and in 1460 the two despots of Morea surrendered. Thomas fled to Italy, Demetrius to the Sultan's court. In 1461 Trebizond, capital of the last remnant of Greek empire, which had maintained its precarious independence by paying court to Turks and Mongols alike, finally succumbed; the transformation of the Byzantine world into the Ottoman world was at last complete.

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(D.M.N.)

Caesar

A Roman general, dictator, and statesman, Gaius Julius Caesar changed the course of the history of the Greco-Roman world decisively and irreversibly. The Greco-Roman society has been extinct for so long that most of the names of its great men mean little to the average, educated modern man. But Caesar's name, like Alexander's, is still on people's lips throughout the Christian and Islāmic worlds. Even people who know nothing of Caesar as a historic personality are familiar with his family name as a title signifying a ruler who is in some sense uniquely supreme or paramount—the meaning of *Kaiser* in German, *tsar* in the Slavonic languages, and *qayṣar* in the languages of the Islāmic world.

Caesar's gens (clan) name, Julius (Iulius), is also familiar in the Christian world; for in Caesar's lifetime the Roman month Quintilis, in which he was born, was renamed "July" in his honour. This name has survived, as has Caesar's reform of the calendar. The old Roman calendar was inaccurate and manipulated for political purposes. Caesar's calendar, the Julian calendar, is still partially in force in the Eastern Orthodox Christian countries; and the Gregorian calendar, now in use in the West, is the Julian, slightly corrected by Pope Gregory XIII.

Caesar, marble bust. In the Capitoline Museum, Rome.
Alinari—Art Resource/EB, Inc.

FAMILY BACKGROUND AND CAREER

Caesar's gens, the Julii, were patricians; *i.e.,* members of Rome's original aristocracy, which had coalesced in the 4th century BC with a number of leading plebeian (commoner) families to form the nobility that had been the governing class in Rome since then. By Caesar's time, the number of surviving patrician gentes was small; and in the gens Julia the Caesares seem to have been the only surviving family. Though some of the most powerful noble families were patrician, patrician blood was no longer a political advantage; it was actually a handicap, since a patrician was debarred from holding the paraconstitutional but powerful office of tribune of the plebs. The Julii Caesares traced their lineage back to the goddess Venus, but the family was not snobbish or conservative-minded. It was also not rich or influential or even distinguished.

A Roman noble won distinction for himself and his family by securing election to a series of public offices, which culminated in the consulship, with the censorship possibly to follow. This was a difficult task for even the ablest and most gifted noble unless he was backed by substantial family wealth and influence. Rome's victory over Carthage in the Second Punic War (218–201 BC) had made Rome the paramount power in the Mediterranean basin; an influential Roman noble family's clients (that is, protégés who, in return, gave their patrons their political support) might include kings and even whole nations, besides numerous private individuals. The requirements and the costs of a Roman political career in Caesar's day were high, and the competition was severe; but the potential profits were of enormous magnitude. One of the perquisites of the praetorship and the consulship was the government of a province, which gave ample opportunity for plunder. The whole Mediterranean world was, in fact, at the mercy of the Roman nobility and of a new class of Roman businessmen, the equites ("knights"), which had grown rich on military contracts and on tax farming.

Military manpower was supplied by the Roman peasantry. This class had been partly dispossessed by an economic revolution following on the devastation caused by the Second Punic War. The Roman governing class had consequently come to be hated and discredited at home and abroad. From 133 onward there had been a series of alternate revolutionary and counter-revolutionary paroxysms. It was evident that the misgovernment of the Roman state and the Greco-Roman world by the Roman nobility could not continue indefinitely and it was fairly clear that the most probable alternative was some form of military dictatorship backed by dispossessed Italian peasants who had turned to long-term military service.

The Julii Caesares

The traditional competition among members of the Roman nobility for office and the spoils of office was thus threatening to turn into a desperate race for seizing autocratic power. The Julii Caesares did not seem to be in the running. It was true that Sextus Caesar, who was perhaps the dictator's uncle, had been one of the consuls for 91; and Lucius Caesar, one of the consuls for 90, was a distant cousin, whose son and namesake was consul for 64. In 90, Rome's Italian allies had seceded from Rome because of the Roman government's obstinate refusal to grant them Roman citizenship, and, as consul, Lucius Caesar had introduced emergency legislation for granting citizenship to the citizens of all Italian ally states that had not taken up arms or that had returned to their allegiance.

Whoever had been consul in this critical year would have had to initiate such legislation, whatever his personal political predilections. There is evidence, however, that the Julii Caesares, though patricians, had already committed themselves to the antinobility party. An aunt of the future dictator had married Gaius Marius, a self-made man (*novus homo*) who had forced his way up to the summit by his military ability and had made the momentous innovation of recruiting his armies from the dispossessed peasants.

The date of Caesar the dictator's birth has long been disputed. The day was July 12 or 13; the traditional (and perhaps most probable) year is 100; but if this date is correct, Caesar must have held each of his offices two years in advance of the legal minimum age. His father, Gaius Caesar, died when Caesar was but 16; his mother, Aurelia, was a notable woman, and it seems certain that he owed much to her.

In spite of the inadequacy of his resources, Caesar seems to have chosen a political career as a matter of course. From the beginning, he probably privately aimed at winning office, not just for the sake of the honours but in order to achieve the power to put the misgoverned Roman state and Greco-Roman world into better order in accordance with ideas of his own. It is improbable that Caesar deliberately sought monarchical power until after he had crossed the Rubicon in 49, though sufficient power to impose his will, as he was determined to do, proved to mean monarchical power.

In 84 Caesar committed himself publicly to the radical side by marrying Cornelia, a daughter of Lucius Cornelius

Sulla

Cinna, a noble who was Marius' associate in revolution. In 83 Lucius Cornelius Sulla returned to Italy from the East and led the successful counter-revolution of 83–82; Sulla then ordered Caesar to divorce Cornelia. Caesar refused and came close to losing not only his property (such as it was) but his life as well. He found it advisable to remove himself from Italy and to do military service, first in the province of Asia and then in Cilicia.

In 78, after Sulla's death, he returned to Rome and started on his political career in the conventional way, by acting as a prosecuting advocate—of course, in his case, against prominent Sullan counter-revolutionaries. His first target, Gnaeus Cornelius Dolabella, was defended by Quintus Hortensius, the leading advocate of the day, and was acquitted by the extortion-court jury, composed exclusively of senators.

Caesar then went to Rhodes to study oratory under a famous professor, Molon. En route he was captured by pirates (one of the symptoms of the anarchy into which the Roman nobility had allowed the Mediterranean world to fall). Caesar raised his ransom, raised a naval force, captured his captors, and had them crucified—all this as a private individual holding no public office. In 74, when Mithradates VI Eupator, king of Pontus, renewed war on the Romans, Caesar raised a private army to combat him.

In his absence from Rome, Caesar was made a member of the politico-ecclesiastical college of pontifices; and on his return he gained one of the elective military tribuneships. Caesar now worked to undo the Sullan constitution in cooperation with Pompey (Gnaeus Pompeius), who had started his career as a lieutenant of Sulla but had changed sides since Sulla's death. In 69 or 68 Caesar was elected quaestor (the first rung on the Roman political ladder). In the same year his wife, Cornelia, and his aunt Julia, Marius' widow, died; in public funeral orations in their honour, Caesar found opportunities for praising Cinna and Marius. Caesar afterward married Pompeia, a distant relative of Pompey. Caesar served his quaestorship in the province of Farther Spain (modern Andalusia and Portugal).

Caesar as political figure

Caesar was elected one of the curule aediles for 65, and he celebrated his tenure of this office by unusually lavish expenditure with borrowed money. He was elected pontifex maximus in 63 by a political dodge. By now he had become a controversial political figure. After the suppression of Catiline's conspiracy in 63, Caesar, as well as the millionaire Marcus Licinius Crassus, was accused of complicity. It seems unlikely that either of them had committed himself to Catiline; but Caesar proposed in the Senate a more merciful alternative to the death penalty, which the consul Cicero was asking for the arrested conspirators. In the uproar in the Senate, Caesar's motion was defeated.

Caesar was elected a praetor for 62. Toward the end of the year of his praetorship, a scandal was caused by Publius Clodius in Caesar's house at the celebration there of the rites, for women only, of Bona Dea (a Roman deity of fruitfulness, both in the Earth and in women). Caesar consequently divorced Pompeia. He obtained the governorship of Farther Spain for 61–60. His creditors did not let him leave Rome until Crassus had gone bail for a quarter of his debts; but a military expedition beyond the northwest frontier of his province enabled Caesar to win loot for himself as well as for his soldiers, with a balance left over for the treasury. This partial financial recovery enabled him, after his return to Rome in 60, to stand for the consulship for 59.

The first triumvirate and the conquest of Gaul. The value of the consulship lay in the lucrative provincial governorship to which it would normally lead. On the eve of the consular elections for 59, the Senate sought to allot to the two future consuls for 59, as their proconsular provinces, the unprofitable supervision of forests and cattle trails in Italy. The Senate also secured by massive bribery the election of an anti-Caesarean, Marcus Calpurnius Bibulus. But they failed to prevent Caesar's election as the other consul.

Caesar now succeeded in organizing an irresistible coalition of political bosses. Pompey had carried out his mission to put the East in order with notable success, but after his return to Italy and his disbandment of his army in 62, the Senate had thwarted him—particularly by preventing him from securing land allotments for his veterans. Caesar, who had assiduously cultivated Pompey's friendship, now entered into a secret pact with him. Caesar's master stroke was to persuade Crassus to join the partnership, the so-called first triumvirate. Crassus—like Pompey, a former lieutenant of Sulla—had been one of the most active of Pompey's obstructors so far. Only Caesar, on good terms with both, was in a position to reconcile them. Early in 59, Pompey sealed his alliance with Caesar by marrying Caesar's only child, Julia. Caesar married Calpurnia, daughter of Lucius Piso, who became consul in 58.

As consul, Caesar introduced a bill for the allotment of Roman public lands in Italy, on which the first charge was to be a provision for Pompey's soldiers. The bill was vetoed by three tribunes of the plebs, and Caesar's colleague Bibulus announced his intention of preventing the transaction of public business by watching the skies for portents whenever the public assembly was convened. Caesar then cowed the opposition by employing some of Pompey's veterans to make a riot, and the distribution was carried out. Pompey's settlement of the East was ratified *en bloc* by an act negotiated by an agent of Caesar, the tribune of the plebs Publius Vatinius. Caesar himself initiated a noncontroversial and much-needed act for punishing misconduct by governors of provinces.

Another act negotiated by Vatinius gave Caesar Cisalpine Gaul (between the Alps, the Apennines, and the Adriatic) and Illyricum. His tenure was to last until February 28, 54. When the governor-designate of Transalpine Gaul suddenly died, this province, also, was assigned to Caesar at Pompey's instance. Cisalpine Gaul gave Caesar a military recruiting ground; Transalpine Gaul gave him a springboard for conquests beyond Rome's northwest frontier.

Caesar's Roman strategy

Between 58 and 50, Caesar conquered the rest of Gaul up to the left bank of the Rhine and subjugated it so effectively that it remained passive under Roman rule throughout the Roman civil wars between 49 and 31. This achievement was all the more amazing in light of the fact that the Romans did not possess any great superiority in military equipment over the north European barbarians. Indeed, the Gallic cavalry was probably superior to the Roman, horseman for horseman. Rome's military superiority lay in its mastery of strategy, tactics, discipline, and military engineering. In Gaul, Rome also had the advantage of being able to deal separately with dozens of relatively small, independent, and uncooperative states. Caesar conquered these piecemeal, and the concerted attempt made by a number of them in 52 to shake off the Roman yoke came too late.

Great though this achievement was, its relative importance in Caesar's career and in Roman history has been overestimated in Western tradition (as have his brief raids on Britain). In Caesar's mind his conquest of Gaul was probably carried out only as a means to his ultimate end. He was acquiring the military manpower, the plunder, and the prestige that he needed to secure a free hand for the prosecution of the task of reorganizing the Roman state and the rest of the Greco-Roman world. This final achievement of Caesar's looms much larger than his conquest of Gaul, when it is viewed in the wider setting of world history and not just in the narrower setting of the Greco-Roman civilization's present daughter civilization in the West.

In 58 Rome's northwestern frontier, established in 125, ran from the Alps down the left bank of the upper Rhône River to the Pyrenees, skirting the southeastern foot of the Cévennes and including the upper basin of the Garonne River without reaching the Gallic shore of the Atlantic. In 58 Caesar intervened beyond this line, first to drive back the Helvetii, who had been migrating westward from their home in what is now central Switzerland. He then crushed Ariovistus, a German soldier of fortune from beyond the Rhine. In 57 Caesar subdued the distant and warlike Belgic group of Gallic peoples in the north, while his lieutenant Publius Licinius Crassus subdued what are now the regions of Normandy and Brittany.

In 56 the Veneti, in what is now southern Brittany, started a revolt in the northwest that was supported by the still unconquered Morini on the Gallic coast of the Straits of Dover and the Menapii along the south bank of the lower Rhine. Caesar reconquered the Veneti with some difficulty and treated them barbarously. He could not finish off the conquest of the Morini and Menapii before the end of the campaigning season of 56; and in the winter of 56–55 the Menapii were temporarily expelled from their home by two immigrant German peoples, the Usipetes and Tencteri. These peoples were exterminated by Caesar in 55. In the same year he bridged the Rhine just below Koblenz to raid Germany on the other side of the river, and then crossed the Channel to raid Britain. In 54 he raided Britain again and subdued a serious revolt in northeastern Gaul. In 53 he subdued further revolts in Gaul and bridged the Rhine again for a second raid.

Vercingetorix

The crisis of Caesar's Gallic war came in 52. The peoples of central Gaul found a national leader in the Arvernian Vercingetorix. They planned to cut off the Roman forces from Caesar, who had been wintering on the other side of the Alps. They even attempted to invade the western end of the old Roman province of Gallia Transalpina. Vercingetorix wanted to avoid pitched battles and sieges and to defeat the Romans by cutting off their supplies—partly by cavalry operations and partly by "scorched earth"—but he could not persuade his countrymen to adopt this painful policy wholeheartedly.

The Bituriges insisted on standing siege in their town Avaricum (Bourges), and Vercingetorix was unable to save it from being taken by storm within one month. Caesar then besieged Vercingetorix in Gergovia near modern Clermont-Ferrand. A Roman attempt to storm Gergovia was repulsed and resulted in heavy Roman losses—the first outright defeat that Caesar had suffered in Gaul. Caesar then defeated an attack on the Roman army on the march and was thus able to besiege Vercingetorix in Alesia, to the northwest of Dijon. Alesia, like Gergovia, was a position of great natural strength, and a large Gallic army came to relieve it; but this army was repulsed and dispersed by Caesar, and Vercingetorix then capitulated.

During the winter of 52–51 and the campaigning season of 51, Caesar crushed a number of sporadic further revolts. The most determined of these rebels were the Bellovaci, between the Rivers Seine and Somme, around Beauvais. Another rebel force stood siege in the south in the natural fortress of Uxellodunum (perhaps the Puy d'Issolu on the Dordogne) until its water supply gave out. Caesar had the survivors' hands cut off. He spent the year 50 in organizing the newly conquered territory. After that, he was ready to settle his accounts with his opponents at home.

Antecedents and outcome of the civil war of 49–45. During his conquest of Gaul, Caesar had been equally busy in preserving and improving his position at home. He used part of his growing wealth from Gallic loot to hire political agents in Rome.

Dissolution of the triumvirate

Meanwhile the cohesion of the triumvirate had been placed under strain. Pompey had soon become restive toward his alarmingly successful ally Caesar, as had Crassus toward his old enemy Pompey. The alliance was patched up in April 56 at a conference at Luca (Lucca), just inside Caesar's province of Cisalpine Gaul. It was arranged that Pompey and Crassus were to be the consuls for 55 and were to get laws promulgated prolonging Caesar's provincial commands for another five years and giving Crassus a five-year term in Syria and Pompey a five-year term in Spain. These laws were duly passed. Crassus was then eliminated by an annihilating defeat at the Parthians' hands in 53. The marriage link between Pompey and Caesar had been broken by Julia's death in 54. After this, Pompey irresolutely veered further and further away from Caesar, until, when the breach finally came, Pompey found himself committed to the nobility's side, though he and the nobility never trusted each other.

The issue was whether there should or should not be an interval between the date at which Caesar was to resign his provincial governorships and, therewith, the command over his armies and the date at which he would enter his proposed second consulship. If there were to be an interval, Caesar would be a private person during that time, vulnerable to attack by his enemies; if prosecuted and convicted, he would be ruined politically and might possibly lose his life. Caesar had to make sure that, until his entry on his second consulship, he should continue to hold at least one province with the military force to guarantee his security.

This issue had already been the object of a series of political manoeuvres and countermanoeuvres at Rome. The dates on which the issue turned are all in doubt. As had been agreed at Luca in 56, Caesar's commands had been prolonged for five years, apparently until February 28, 49, but this is not certain. In 52, a year in which Pompey was elected sole consul and given a five-year provincial command in Spain, Caesar was allowed by a law sponsored by all 10 tribunes to stand for the consulship *in absentia*. If he were to stand in 49 for the consulship for 48, he would be out of office, and therefore in danger, during the last 10 months of 49. As a safeguard for Caesar against this, there seems to have been an understanding—possibly a private one at Luca in 56 between him and Pompey—that the question of a successor to Caesar in his commands should not be raised in the Senate before March 1, 50. This manoeuvre would have ensured that Caesar would retain his commands until the end of 49. However, the question of replacing Caesar was actually raised in the Senate a number of times from 51 onward; each time Caesar had the dangerous proposals vetoed by tribunes of the plebs who were his agents—particularly Gaius Scribonius Curio in 50 and Mark Antony in 49.

The issue was brought to a head by one of the consuls for 50, Gaius Claudius Marcellus. He obtained resolutions from the Senate that Caesar should lay down his command (presumably at its terminal date) but that Pompey should not lay down his command simultaneously. Curio then obtained on December 1, 50, a resolution (by 370 votes to 22) that both men should lay down their commands simultaneously. Next day Marcellus (without authorization from the Senate) offered the command over all troops in Italy to Pompey, together with the power to raise more; and Pompey accepted. On January 1, 49, the Senate received from Caesar a proposal that he and Pompey should lay down their commands simultaneously. Caesar's message was peremptory, and the Senate resolved that Caesar should be treated as a public enemy if he did not lay down his command "by a date to be fixed."

Caesar crosses the Rubicon

On January 10–11, 49, Caesar led his troops across the little river Rubicon, the boundary between his province of Cisalpine Gaul and Italy proper. He thus committed the first act of war. This was not, however, the heart of the matter. The actual question of substance was whether the misgovernment of the Greco-Roman world by the Roman nobility should be allowed to continue or whether it should be replaced by an autocratic regime. Either alternative would result in a disastrous civil war. The subsequent partial recuperation of the Greco-Roman world under the principate suggests, however, that Caesarism was the lesser evil.

The civil war was a tragedy, for war was not wanted either by Caesar or by Pompey or even by a considerable part of the nobility, while the bulk of the Roman citizen body ardently hoped for the preservation of peace. By this time, however, the three parties that counted politically were all entrapped. Caesar's success in building up his political power had made the champions of the old regime so implacably hostile to him that he was now faced with a choice between putting himself at his enemies' mercy or seizing the monopoly of power at which he was accused of aiming. He found that he could not extricate himself from this dilemma by reducing his demands, as he eventually did, to the absolute minimum required for his security. As for Pompey, his growing jealousy of Caesar had led him so far toward the nobility that he could not come to terms with Caesar again without loss of face.

The first bout of the civil war moved swiftly. In 49 Caesar drove his opponents out of Italy to the eastern side of the Straits of Otranto. He then crushed Pompey's army in Spain. Toward the end of 49, he followed Pompey across the Adriatic and retrieved a reverse at Dyrrachium by winning a decisive victory at Pharsalus on August 9, 48. Caesar pursued Pompey from Thessaly to Egypt, where Pompey was murdered by an officer of King Ptolemy. Caesar wintered in Alexandria, fighting with the populace and dallying with Queen Cleopatra. In 47 he fought a brief local war in northeastern Anatolia with Pharnaces, king of the Cimmerian Bosporus, who was trying to regain his father Mithradates' kingdom of Pontus. Caesar's famous words, *Veni, vidi, vici* ("I came, I saw, I conquered"), are his own account of this campaign.

Caesar then returned to Rome, but a few months later, now with the title of dictator, he left for Africa, where his opponents had rallied. In 46 he crushed their army at Thapsus and returned to Rome, only to leave in November for Farther Spain to deal with a fresh outbreak of resistance, which he crushed on March 17, 45, at Munda. He then returned to Rome to start putting the Greco-Roman world in order. He had less than a year's grace for this huge task of reconstruction before his assassination in the Senate House at Rome on March 15, 44 (the Ides of March).

Caesar's death

Caesar's death was partly due to his clemency and impatience, which, in combination, were dangerous for his personal security. Caesar had not hesitated to commit atrocities against "barbarians" when it had suited him, but he was almost consistently magnanimous in his treatment of his defeated Roman opponents. Thus clemency was probably not just a matter of policy. Caesar's earliest experience in his political career had been Sulla's implacable persecution of his defeated domestic opponents. Caesar amnestied his opponents wholesale and gave a number of them responsible positions in his new regime. Gaius Cassius Longinus, who was the moving spirit in the plot to murder him, and Marcus Junius Brutus, the symbolic embodiment of Roman republicanism, were both former enemies. "*Et tu, Brute*" ("You too, Brutus") was Caesar's expression of his particular anguish at being stabbed by a man whom he had forgiven, trusted, and loved.

There were, however, also a number of ex-Caesareans among the 60 conspirators. They had been goaded into this volte-face by the increasingly monarchical trend of Caesar's regime and, perhaps at least as much, by the aristocratic disdain that inhibited Caesar from taking any trouble to sugar the bitter pill. Some stood to lose, rather than to gain, personally by the removal of the autocrat who had made their political fortunes. But even if they were acting on principle, they were blind to the truth that the reign of the Roman nobility was broken beyond recall and that even Caesar might not have been able to overthrow the *ancien régime* if its destruction had not been long overdue. They also failed to recognize that by making Caesar a martyr they were creating his posthumous political fortune.

If Caesar had not been murdered in 44, he might have lived on for 15 or 20 years. His physical constitution was unusually tough, though in his last years he had several epileptic seizures. What would he have done with this time? The answer can only be guessed from what he did do in the few months available. He found time in the year 46 to reform the Roman calendar. In 45 he enacted a law laying down a standard pattern for the constitutions of the *municipia,* which were by this time the units of local self-government in most of the territory inhabited by Roman citizens. In 59 Caesar had already resurrected the city of Capua, which the republican Roman regime more than 150 years earlier had deprived of its juridical corporate personality; he now resurrected the other two great cities, Carthage and Corinth, that his predecessors had destroyed. This was only a part of what he did to resettle his discharged soldiers and the urban proletariat of Rome. He was also generous in granting Roman citizenship to aliens. (He had given it to all of Cisalpine Gaul, north of the Po, in 49.) He increased the size of the Senate and made its personnel more representative of the whole Roman citizenry.

At his death, Caesar was on the point of starting out on a new military campaign to avenge and retrieve Crassus' disastrous defeat in 53 by the Parthians. Would Caesar have succeeded in recapturing for the Greco-Roman world the extinct Seleucid monarchy's lost dominions east of the Euphrates, particularly Babylonia? The fate of Crassus' army had shown that the terrain in northern Mesopotamia favoured Parthian cavalry against Roman infantry. Would Caesar's military genius have outweighed this handicap? And would Rome's hitherto inexhaustible reservoir of military manpower have sufficed for this additional call upon it? Only guesses are possible, for Caesar's assassination condemned the Romans to another 13 years of civil war, and Rome would never again possess sufficient manpower to conquer and hold Babylonia.

PERSONALITY AND REPUTATION

Caesar was not and is not lovable. His generosity to defeated opponents, magnanimous though it was, did not win their affection. He won his soldiers' devotion by the victories that his intellectual ability, applied to warfare, brought them. Yet, though not lovable, Caesar was and is attractive, indeed fascinating. His political achievement required ability, in effect amounting to genius, in several different fields, including administration and generalship besides the minor arts of wire pulling and propaganda.

In all these, Caesar was a supreme virtuoso. But if he had not also been something more than this he would not have been the supremely great man that he undoubtedly was.

Caesar was great beyond—and even in conflict with—the requirements of his political ambition. He showed a human spiritual greatness in his generosity to defeated opponents, which was partly responsible for his assassination. (The merciless Sulla abdicated and died in his bed.)

Speeches and writings

Another field in which Caesar's genius went far beyond the requirements of his political ambition was his writings. Of these, his speeches, letters, and pamphlets are lost. Only his accounts (both incomplete and supplemented by other hands) of the Gallic War and the civil war survive. Caesar ranked as a masterly public speaker in an age in which he was in competition first with Hortensius and then with Cicero.

All Caesar's speeches and writings, lost and extant, apparently served political purposes. He turned his funeral orations for his wife and for his aunt to account, for political propaganda. His accounts of his wars are subtly contrived to make the unsuspecting reader see Caesar's acts in the light that Caesar chooses. The accounts are written in the form of terse, dry, factual reports that look impersonal and objective, yet every recorded fact has been carefully selected and presented. As for the lost *Anticato,* a reply to Cicero's eulogy of Caesar's dead opponent Marcus Porcius Cato, it is a testimony to Caesar's political insight that he made the time to write it, in spite of the overwhelming military, administrative, and legislative demands on him. He realized that Cato, in giving his life for his cause (46), had made himself posthumously into a much more potent political force than he had ever been in his lifetime. Caesar was right, from his point of view, to try to put salt on Cato's tail. He did not succeed, however. For the next 150 years, Cato the martyr continued to be a nuisance, sometimes a menace, to Caesar's successors.

The mark of Caesar's genius in his writings is that though they were written for propaganda they are nevertheless of outstanding literary merit. A reader who has seen through their prosaic purpose can ignore it and appreciate them as splendid works of art.

Caesar's most amazing characteristic is his energy, intellectual and physical. He prepared his seven books on the Gallic War for publication in 51 when he still had serious revolts in Gaul on his hands, and he wrote his books on the civil war and his *Anticato* in the hectic years between 49 and 44. His physical energy was of the same order. For instance, in the winter of 57–56 he found time to visit his third province, Illyria, as well as Cisalpine Gaul; and in the interval between his campaigns of 55 and 54 he transacted public business in Cisalpine Gaul and went to Illyria to settle accounts with the Pirustae, a turbulent tribe in what is now Albania. In 49 he marched, within a single campaigning season, from the Rubicon to Brundisium and from Brundisium to Spain. At Alexandria, probably aged 53, he saved himself from sudden death by his prowess as a swimmer.

Caesar's physical vitality perhaps partly accounts for his sexual promiscuity, which was out of the ordinary, even by contemporary Greek and Roman standards. It was rumoured that during his first visit to the East he had had homosexual relations with King Nicomedes of Bithynia. The rumour is credible, though not proved, and was repeated throughout Caesar's life. There is no doubt of Caesar's heterosexual affairs, many of them with married women. Probably Caesar looked upon these as trivial recreations. Yet he involved himself at least twice in escapades that might have wrecked his career. If he did in fact have an affair with Pompey's wife, Mucia, he was risking his entente with Pompey. A more notorious, though not quite so hazardous, affair was his liaison with Cleopatra. By dallying with her at Alexandria, he risked losing what he had just won at Pharsalus. By allowing her to visit him in Rome in 46, he flouted public feeling and added to the list of tactless acts that, cumulatively, goaded old comrades and amnestied enemies into assassinating him.

Caesar's achievement

This cool-headed man of genius with an erratic vein of sexual exuberance undoubtedly changed the course of history at the western end of the Old World. By liquidating the scandalous and bankrupt rule of the Roman nobility, he gave the Roman state—and with it the Greco-Roman civilization—a reprieve that lasted for more than 600 years in the East and for more than 400 years in the relatively backward West. Caesar substituted for the Roman oligarchy an autocracy that could never afterward be abolished. If he had not done this when he did it, Rome and the Greco-Roman world might have succumbed, before the beginning of the Christian era, to barbarian invaders in the West and to the Parthian Empire in the East. The prolongation of the life of the Greco-Roman civilization had important historical effects. Under the Roman Empire the Near East was impregnated with Hellenism for six or seven more centuries. But for this the Hellenic element might not have been present in sufficient strength to make its decisive impact on Christianity and Islām. Gaul, too, would have sunk deeper into barbarism when the Franks overran it, if it had not been associated with the civilized Mediterranean world for more than 500 years as a result of Caesar's conquest.

Caesar's political achievement was limited. Its effects were confined to the western end of the Old World and were comparatively short-lived by Chinese or ancient Egyptian standards. The Chinese state founded by Shih Huang Ti in the 3rd century BC still stands, and its future may be still greater than its past. Yet, even if Caesar were to prove to have been of lesser stature than this Chinese colossus, he would still remain a giant by comparison with the common run of human beings (see also GRECO-ROMAN CIVILIZATION: *Ancient Rome*).

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(A.J.T./Ed.)

Cairo

Cairo, the capital of Egypt, is the largest city in Africa. Its Arabic name, al-Qāhirah, means "Victorious." The city has stood for more than 1,000 years on the same site on the banks of the Nile, primarily on the eastern shore, some 500 miles (800 kilometres) downstream from the Aswān High Dam. Cairo is the gateway to the Nile Delta, which begins about 10 miles to the north where the lower Nile separates into the Rosetta and Damietta branches, and it has served for centuries as the stronghold from which to defend all of Egypt to the south.

Cairo is a place of vivid contrasts. Along the well-irrigated shoreline, lush tropical vegetation, tall palms, flowering flame trees, and skyscrapers are profiled against a cloudless sky; in the older inland quarters to the east, beneath the foothills of the Arabian Desert (aṣ-Ṣaḥrā' ash-Sharqīyah; "Eastern Desert") and the rocky promontories of the Muqaṭṭam Hills and the Red Mountain (al-Jabal al-Aḥmar, with its petrified forest), browns and ochres are the dominant hues of land and buildings.

The city juxtaposes ancient and new, East and West. The Great Pyramids near Memphis stand at the southwestern edge of the metropolis, and an obelisk in the northeast marks the site of Heliopolis, where Plato once studied; modern landmarks include elegant Western-style high-rise hotels overlooking the Nile River. Between these extremes are other architectural monuments dating from Roman, Arab, and Turkish times. In addition to department stores, cinemas, hotels, and town houses, Cairo contains a large functioning bazaar and an extensive medieval city endowed with more than 400 registered historic monuments—mosques, mausoleums, crenellated walls, and massive gateways—dating from AD 130 to the early 19th century.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

Climate. The typical desert climate contrasts daytime dry heat with cool nights freshened by Nile breezes. Cairo has only two seasons: approximately eight months of summer and four months of winter. In the hottest of the summer months, June, July, and August, the average daily maximum temperature is 95° F (35° C) and the average daily minimum is 70° F (21° C). The summer temperature has reached as high as 117° F (47° C). During winter, days are warmed by the strong Tropic of Cancer sun, but nights are cool and humid. The average daily maximum temperature in January–February is 67° F (19° C) and the average daily minimum is 47° F (9° C). A healthful climate has long made Cairo a renowned winter resort.

The city layout. *The old and new quarters.* Cairo is fan-shaped, narrowest in the south, where the river valley is wedged between desert escarpments, and widest in the north, where the valley blends into the Delta. Over the centuries the city expanded westward as a receding river channel left land flood-free. In response to heightened demand, however, the city also has been elongated to the north and south and has developed an expanding annex on the Nile's western shore.

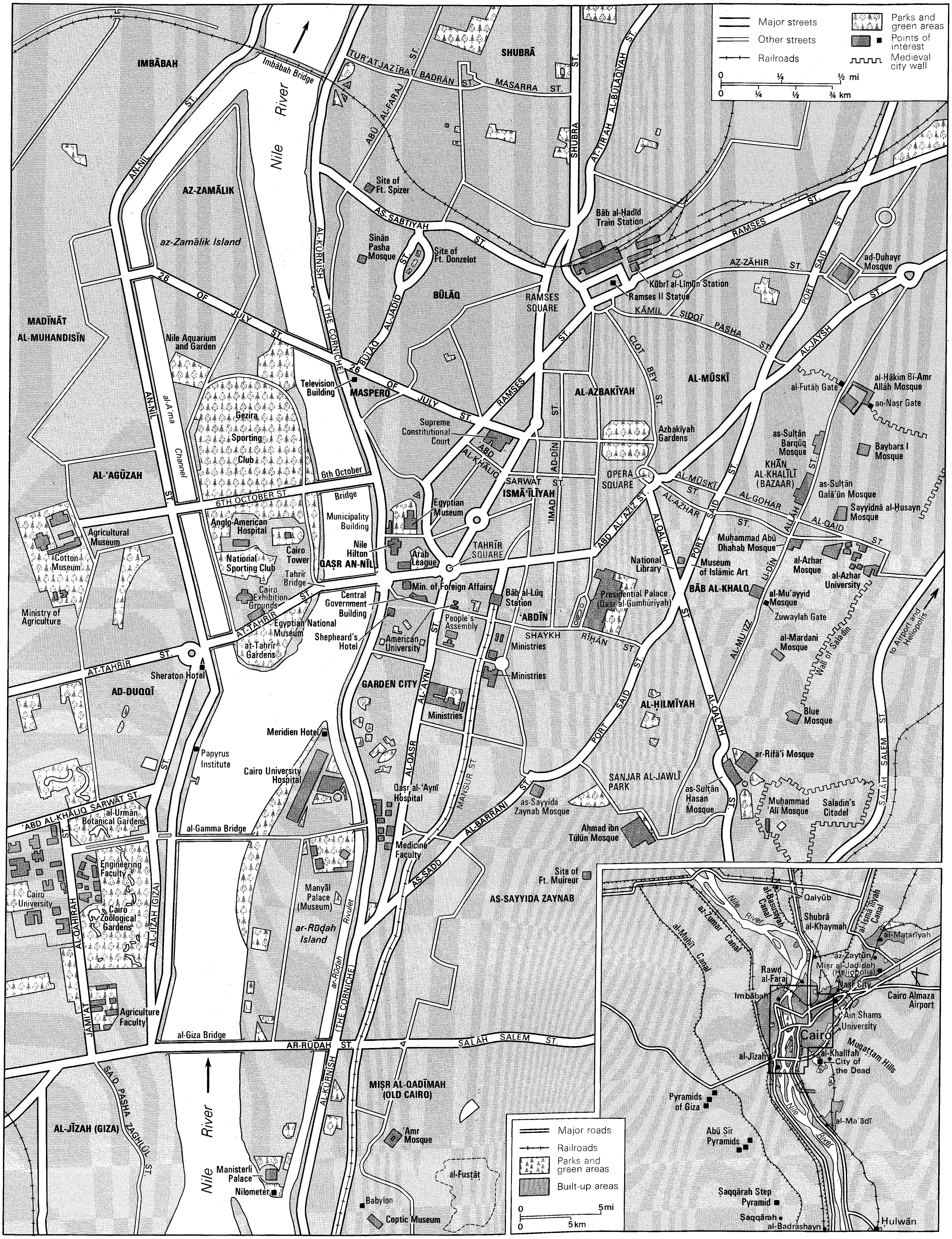
The *muḥāfaẓah* (governorate) of Cairo is one of the administrative districts into which Egypt is divided. The boundaries of the governorate encompass only half of the urbanized area; Ḥulwān, an industrial satellite in the extreme south, is included, but industrial satellites in the far north, such as Shubrā al-Khaymah, and the heavily developed quarters on the west bank belong to other *muḥāfaẓāt*.

Sectors of the city

The organization of the metropolitan complex is understandable only in the context of the city's history. The three oldest areas constitute densely populated slums that virtually surround the Westernized core. The largest of these is the city built under the Fāṭimid dynasty, with pre-19th-century extensions (al-Jamālīyah, ad-Darb al-Aḥmar, Bāb ash-Sha'rīyah, eastern as-Saiyīdah Zīnab, northern al-Khalīfah). In this densely settled zone containing the oldest buildings are most of Cairo's historic monuments, from the Mosque of Baybars I at the northern edge to Saladin's Citadel in the south. Among the major bazaars near al-Azhar Mosque in the central walled city are the Khān al-Khalīlī (1390) and the markets for gold articles, copper ware, textiles, rugs, amber, spices, and leather goods. Two major thoroughfares run north and south: Mu'izz li-Dīn Allāh, which bisects the old city and contains the major mosques and markets; and Port Said Street, which runs along the bed of an ancient canal that once marked the western border of Fāṭimid Cairo. Al-Gohar al-Qaid Street (which runs into al-Mūskī Street) and al-Azhar Street connect the old city with the central business district. A diagonal, al-Qal'ah Street, links the central business district with the Citadel. Most other streets are narrow, twisting, and often dead-end. Only major thoroughfares are open to motor vehicles.

The other two old quarters are Būlāq, to the northwest of the medieval city, and Miṣr al-Qadīmah (Old Cairo), to the south; they served as port suburbs of Cairo before the city expanded to encompass them. Būlāq, an island until 1340 and the main port by 1560, became an industrial quarter in the early 19th century. It contains small workshops, the National Press, textile factories, and trade schools. A poor population, including many village migrants, is housed there at extremely high densities. The mosques of Abū al-'Alā' and Sinān Pasha are among the few historic buildings remaining; many of the older structures have been razed to make room for commercial and office buildings. Miṣr al-Qadīmah is an even poorer quarter, although only a small section is ancient. The walled compound of Babylon, with its semi-submerged Roman bastion, its Coptic churches and museum, and its ancient synagogue and houses, is virtually intact, as is the reconstructed 'Amr Mosque nearby. The rich agricultural land between Cairo and Ḥulwān has been given over to apartments, and the residential suburb of al-Ma'ādī, built for single-family residences under the British, now hosts several skyscrapers.

Flanked by these older quarters are the central business district (al-Azbakīyah) and its residential quarters (Garden



Cairo and (inset) its metropolitan area.

City, Ismāʿīlīyah), which spread onto the Gezīrah, the island offshore. The major thoroughfare is al-Kūrnīsh (the Corniche), a highway paralleling the river, built since the overthrow of the monarchy in 1952. From north to south are the imposing Television Building, the Ramses Hilton Hotel, the Municipality Building, the Nile Hilton Hotel (with the Egyptian Museum behind it), and Shepheard's Hotel. Also in the vicinity are the intricately curved streets of Garden City, lined with tall apartment houses. The Gezīrah, across the Nile from al-Kūrnīsh, is the site of hotels, the Cairo Tower, museums, public gardens, a racetrack, and sporting and officers' clubs, as well as elegant housing.

Between the imposing Nile riverfront and the ancient inland quarters is a transitional zone (al-Mūskī, Bāb al-Lūq, eastern ʿAbdīn, an-Naṣrīyah, al-Ḥilmīyah) of working-class status. It was chiefly developed in the 19th century and contains the National Library, the Museum of Islāmic Art, and the Presidential Palace (Qaṣr al-Gumhūriyah) and archives.

The City
f the
ead"

Along the eastern and southern edges of the metropolis are extensive cemeteries—a zone called "the City of the Dead," which has no counterpart outside Egypt. In a vast, dusty, ochre-coloured zone unserved by municipal utilities are found exquisite shrine-mosques and mausoleums of early religious leaders (such as that of Imām ash-Shāfiʿī, an al-Fusṭāṭ resident and founder of Egypt's major legal tradition) and of Mamlūk sultans (among the most beautiful are the memorials of Qāʾit Bāy, Barqūq, and Qalāʾūn). More modest and modern structures are also found there. Particularly since the population increase that occurred during World War II, housing and shops have been added to the City of the Dead. Some 250,000 Cairenes live there, for the most part without government sanction.

The northern and western peripheries have grown dramatically. On the west bank and on the island of ar-Rūḍah are located developed residential quarters (ad-Duqqī and al-Jīzah, or Giza), the zoological and botanical gardens, an agricultural museum, and the campus of Cairo University. A large-scale housing project is Workers' City in Imbābah, while Engineers' City (Madīnāt al-Muhandisīn) is now the domain of Cairo's middle and upper middle classes. Expansion has also taken place to the north. Beginning about 1905 but expanding substantially in the 1920s, the northern quarters of Rawḍ al-Faraj, Shubrā, Sharabīyah, al-Qubbah, al-ʿAbbāsīyah, al-Maṭarīyah, az-Zaytūn, and Miṣr al-Jadīdah (Heliopolis) gained population. Since that time urban developments have increasingly encroached on agricultural land and have been extended into the desert periphery by elaborate irrigation schemes. Heliopolis, first conceived in 1905, and Naṣr City, a new town begun in 1958, are examples of such developments. A rural population still inhabits the northern fringe, but squatter settlements are also found there. The newest zones of Cairo's metropolitan area, however, are a series of satellite towns being built in desert areas beyond the city centre. Sites and services housing projects, in which the government designs the subdivision and puts in streets, sewers, electricity, and water lines, also occupy desert peripheries.

Architectural styles. Building styles in Cairo are related to the historic period during which each quarter developed. In the oldest sections, two- to four-storied structures prevail, most built of fired brick covered with plaster and sometimes shored with half-timbering. Wood, being scarce, is used frugally. Some of the oldest homes have windows covered with delicately turned wooden lattices (*mashrabīyah*) and massive wooden doors decorated with inlay, brass, or iron nailheads, indicating past elegance. The traditional dwellings (of which only a few remain intact) open onto fountained courtyards and have separate quarters for men and women; the traditional workshops and warehouse inns (of which more have survived) have galleries overlooking the interior court.

European
architectural
influences

Parts of Cairo built in the 19th century reveal exaggerated European influences—highly ornate stone exteriors, cupolas, and Romanesque doorways. While this ungainly and incongruous style, darkened with time, predominates in the transitional zone, perhaps the most outlandish examples are the later Palace of Sakākīnī (northeast of Baybars I Mosque) and the 20th-century palace of Baron Empain, founder of Heliopolis.

In the early modern Western quarters, built at the turn of the 20th century, the architectural style is Parisian; most of the moderately tall buildings are of stone or poured concrete. Closer to the Nile and on the islands, a contemporary Mediterranean style predominates. Tall reinforced concrete and glass structures with balconies are decorated with tile. In these less derivative forms, Egyptian architectural genius, so noticeable in the clean, almost stark lines of the ancient temples, seems to be reemerging.

The monuments of the eastern cemeteries are Mamlūk in design, each topped by a plain or fluted dome; the lesser tombs are simpler rectangles. Houses there and in the rural fringe areas typically are built of mud or of crudely fired brick, resembling traditional village housing in the hinterland.

Robert Frerck/Odyssey Productions

Al-Azhar Mosque (domed building on right), with adjoining buildings of al-Azhar University (founded in 970).

In the quarters on the west and north the more elegant districts have both handsome high-rise apartments and one- or two-storied "villas," with high walls enclosing colourful gardens. Lower middle- and working-class housing consists exclusively of concrete multiflat structures, gray or yellow-beige in colour, often with shops occupying the ground floor. The poorest zones contain three- to five-storied walk-ups, often of crudely fired redbrick.

All roofs are flat. In most quarters commercial buildings are scattered among the dwellings, and in the poorer quarters one also finds an occasional industrial workshop.

THE PEOPLE

Cairo's population previously was diverse, both ethnically and religiously, but the vast majority now is Muslim. The remainder are Egyptian Christians of the Coptic Orthodox faith. Members of the once-dominant foreign groups (English, French, Swiss, and Belgian) had mostly left by 1957; the more modest Italian and Greek communities remained and have been joined by Arab nationals (Syrians, Palestinians, and Sudanese) as well as by diplomatic and technical personnel from eastern Europe. Differences of status within the Egyptian population depend largely on place of origin (many residents of Cairo were born in rural Egypt), class, and degree of modernity. About half of the population lives in the city proper, and half lives in the suburbs.

THE ECONOMY

Industry. From its inception the economy of Cairo has been based on governmental functions, commerce, trade, and industrial production. Especially since the 1920s, the modern productive sector has expanded dramatically. Large-scale industrialization since the 1952 revolution has built upon previous developments in textiles (utilizing the long-staple cotton for which Egypt is famous) and food processing, which consists of canning and freezing the wide variety of fruits and vegetables grown in the fertile Delta, as well as processing the tobacco and sugarcane grown in Upper Egypt. In addition to the production of iron and steel, consumer goods, such as cars and refrigerators, are made. Especially since 1956, the output of a wide variety of smaller consumer goods has increased.

Cairo's significance within the nation

Cairo is the primary national centre for modern economic production and financial control. It contains most of Egypt's important banks, shipping companies, and airlines; about one-third of all sports, amusement, cafés, restaurants, and hotels in the country are also concentrated in Cairo.

Although about one-half of Cairo's predominantly male labour force works in the expanding modern sector of the economy, both on the assembly lines and in auxiliary commercial and financial institutions, a substantial traditional sector still survives, in which craftsmanship and personal relationships play an important role.

Labour. About one-third of all workers are engaged in service occupations, many of which are of only dubious productivity and viability. The government's development plans are directed toward mobilizing this labour force. Unemployment is relatively high in Cairo. In the mid-1980s as many as 3,200,000 Egyptian labourers were employed as guest workers, primarily in the oil-producing Persian Gulf states and Libya, but many were repatriated when oil prices fell and production slowed. Their return has helped to drive Cairo's unemployment even higher.

Transportation. Cairo is served by an international airport beyond Heliopolis and by train service to other major cities. Air-conditioned trains connect Alexandria with the Bāb al-Ḥadīd Train Station in Cairo, and there is overnight sleeper service to Luxor and Aswān. The city itself has mass-transit facilities. Construction began in 1981 on an extensive metropolitan rail system, and its first phase, which concentrated on the upgrading of the existing line between Cairo and Ḥulwān to the south, was completed in 1987. The hub of the metropolitan system is a 2.7-mile subway network under central Cairo, construction of which began in 1982. Streetcars and buses thread through all but the Fāṭimid quarters. Local assembly plants as well as imports have made an increasing number of automobiles available, and bypass highways and overpasses have been built to relieve the growing traffic congestion. In the old city, donkey-drawn carts are still seen, but they are rapidly disappearing.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Before 1949 Cairo was anomalously administered by a combination of national policymakers and local administrators. In 1949 the municipality (*baladīyah*) of Cairo was created, being inaugurated, together with a town council, in 1950. In 1960 the national Ministry for Local Administration was established, after which it promulgated a uniform system for local governments. Although the ministry was dissolved in 1971, some basic changes introduced have been retained, including the merger of the *baladīyah* with the *muḥāfaẓah* (governorate) of Cairo.

At the head of the governorate of Cairo is the governor, who is appointed by the president and assisted by the executive committee of the governorate. This committee includes under secretaries from the major national ministries, such as education, housing, health, social affairs, finance, and the interior. Several special administrations are also under the committee's direction. The Municipal Council of Cairo (Majlis Muḥāfazah al-Qāhirah), consisting of both elected representatives and members ex officio, was dissolved in 1971 and replaced by the Popular Assembly (al-Majlis al-Sha'abī), in which only elected representatives can vote. The central government still maintains financial control over local programs and budgets.

Public utilities. Gas and water systems have existed since the 1860s and electrical and sewerage systems since the early 20th century. The majority of the city's older dwellings are connected to the electrical grid, but fewer than one-half have water and sewer connections. The major challenge has been to extend utility lines of all types to new quarters and suburbs and to repair and maintain the older installations. Nile water is filtered and purified, and tap water is both safe and pleasant to drink. The sewerage system has been substantially renovated and its capacity enlarged. Telephone service was substantially renovated in 1985.

Health. Most of Egypt's health and medical facilities are concentrated in Cairo. There are several government hospitals, including the enormous general hospitals of Qaṣr al-ʿAynī and Dimardash and the mental hospital at ʿAbbāssīyah. The city is also served by numerous smaller private hospitals (such as Dār al-Shifāʾ, the Italian Hospital, and the Anglo-American Hospital). Special clinics and various public hospitals are devoted to prevalent eye diseases (chiefly trachoma).

Education. Primary education is compulsory, and the illiteracy rate has been greatly reduced. In addition to primary and secondary schools, the educational facilities include technical institutes, pre-university superior schools, and foreign schools. Both the number of schools and the number of pupils have risen markedly.

Higher education

Cairo is also Egypt's centre of higher education and attracts not only Egyptian students but also students from other Arab countries. The various faculties of Cairo University produce the country's largest number of college graduates and specialized professionals (doctors, lawyers, and engineers). ʿAin Shams University is also a notable institution, and al-Azhar University, which previously specialized in religious subjects and language and literature, now includes teaching in the sciences on its supplementary campus in Naṣr City. The privately run American University in Cairo offers instruction in English in various arts and sciences.

CULTURAL LIFE

Cairo is the cultural capital of the Middle East as well as the region's chief mass media centre. For many centuries the region's major religious and cultural institutions have been concentrated there. During the 19th century European cultural institutions, such as theatre and opera, were added. The Baroque Opera House, where music by Egyptian composers was sometimes heard, was home to the Cairo Symphony Orchestra and host to guest ballet and opera companies, including the indigenous Reḍā Folklore Ballet Troop. The opera house, on Opera Square in the downtown area, was destroyed by fire in 1971, and construction of a replacement on the southern tip of the Gezīrah was undertaken in 1986.

Egypt has long been known for its musical and dramatic talent, and there has been a renaissance of the legitimate Arabic theatre. The majority of Arabic films are produced by Egyptian companies in Cairo. The leading cinema stars, as well as the most popular musical entertainers of the Arab world, make Cairo their headquarters. Egyptian radio is broadcast throughout North Africa, into the Fertile Crescent, and south into Africa.

Cairo is rich in museums, such as the Egyptian Museum, on Taḥrīr Square, which displays the treasures of Tutankhamen; the Coptic Museum in Miṣr al-Qadīmah, specializing in pre-Islāmic icons, textiles, and stones; the beautifully renovated Museum of Islāmic Art in Bāb al-Khalq, with Mamlūk Qur'āns and objects of wood, brass, inlay, and glass; the War (Citadel) Museum; and the Turkish-style Manyāl Palace Museum (on ar-Rūḍah Island). The mosques of Cairo offer as rich a store as any museum.

The Cairo Zoological Gardens in al-Jīzah contain extensive collections of rare tropical animals in a garden setting. Entertainment facilities available include sailboat trips up the Nile, as well as innumerable riverfront cafés, restaurant boats, and nightclubs. In addition to the older sporting facilities on the Gezīrah (racetracks, swimming clubs, and gardens), there is a racetrack at Heliopolis. Naṣr City is the site of the Cairo Stadium and has numerous playing fields.

History

THE EARLY PERIOD

Although ruined Memphis, 14 miles southwest of Cairo, was a metropolis 5,000 years ago, and about 2,000 years ago the Romans occupied a town on the site of Cairo called Babylon (now the Miṣr al-Qadīmah quarter), the seed from which contemporary Cairo sprang was the town of al-Fusṭāṭ, founded as a military encampment in AD 641 by 'Amr ibn al-'Āṣ, commander of the Arabs who brought Islām to Egypt. Successor dynasties added royal suburbs (al-'Askar, founded in 750 by the Umayyads; al-Qaṭā'i', founded in 870 by Aḥmad ibn Ṭūlūn) to the increasingly prosperous commercial and industrial port city of al-Fusṭāṭ. Little remains of these early developments in the southern part of the city, except the tower of Trajan (AD 130), the mosques of 'Amr ibn al-'Āṣ (641) and Aḥmad ibn Ṭūlūn (878), and the partially excavated mounds covering the site of al-Fusṭāṭ.

In 969 adherents of a dissident Islāmic sect, the Fāṭimids, invaded Egypt from what is now Tunisia. The conquering general, Jawhar, established a new rectangular walled city northeast of existing settlements. Initially named al-Manṣūrīyah, the city was renamed al-Qāhirah in 973–974 when the Fāṭimid caliph al-Mu'izz arrived to make it the capital of a dynasty that lasted for 200 years. Al-Qāhirah and al-Fusṭāṭ coexisted until 1168, when unfortified al-Fusṭāṭ was set on fire to protect Cairo from the crusaders. The crusaders were driven off by a Sunnī (orthodox Islāmic) army from Syria, after which the victorious commander, Saladin, founded the Ayyūbid dynasty, which controlled a vast empire from Cairo.

The Mamlūk era

Even though al-Fusṭāṭ was partially rebuilt, Cairo itself became transformed from a royal enclave into an imperial metropolis. Saladin further extended the 11th-century walls built by a high official called Badr al-Jamālī (the northern and southern walls and three main gates, al-Futuḥ Gate, an-Naṣr Gate, and Zuwaylah Gate, are still extant) and constructed a citadel on the Muqaṭṭam spur (now dominated by the Muḥammad 'Alī Mosque). After 1260, when Baybars I became the first Mamlūk sultan of undisputed legitimacy, Cairo served as the capital of the Mamlūk empire, which governed Egypt and the Fertile Crescent until 1516.

Medieval Cairo reached its apogee during the Mamlūk era. By about 1340, almost 500,000 persons lived in an area five times greater than the original Fāṭimid walled city, and Cairo had become the greatest city of Africa, Europe, and Asia Minor. Its al-Azhar University was the principal seat of Islāmic learning. The city was a key link in the profitable East–West spice trade and the recipient of tribute from a wealthy empire. Most of Cairo's greatest architectural masterpieces were built during this epoch.

Decline set in thereafter—sporadically at first, and then precipitously. The population was decimated by plagues, including the Black Death in 1348. The spice trade monopoly was broken by Vasco da Gama's voyage from Portugal to India (1497–99). Finally, political autonomy was lost to the conquering Turks, who, after 1517, reduced Cairo to a provincial capital. In 1798, when Napoleon and his troops arrived in Cairo, fewer than 300,000 people were living in the city and its two port suburbs, Miṣr al-Qadīmah and Būlāq. The Turks returned after Napoleon's defeat in 1801. In 1805 Muḥammad 'Ali, commander of an Albanian contingent, was appointed pasha, thus founding the dynasty that ruled Egypt until his great-great-grandson, Farouk I, abdicated in 1952.

DEVELOPMENT OF THE CITY

Modern urban growth began in the 1830s, but only during Ismā'īl's reign (1863–79) was the city fundamentally transformed. Influenced by Baron Haussmann's renovation of Paris, Ismā'īl ordered the construction of a European-style city to the west of the medieval core. French city-planning methods dominated the design of the districts of al-Azbakīyah (with its large park), 'Abdīn, and Ismā'īlīyah—all now central zones of contemporary Cairo. By the end of the 19th century these districts were well developed, but with the rise of British hegemony from 1882 onward, they became transformed into a colonial enclave. Following the overthrow of the monarchy in 1952, however, the number and power of foreign residents declined.

The 20th century

During the 20th century, Cairo grew spectacularly in both population and area. Improvements in transportation fostered the growth of suburban Heliopolis and al-Ma'ādī; flood control permitted riverfront development; bridges encouraged settlement of islands (ar-Rūḍah and az-Zamālik) and of the west bank. By mid-century the city was primarily growing northward into the fertile Delta, a trend further encouraged by industrialization. Since 1952 the government has attempted to accommodate Cairo's rapidly increasing population by building planned suburbs (Naṣr City, Muqattam City, Engineers' City) and, more recently, distant satellite towns, such as May 15th, October 6th, and Tenth of Ramaḍān towns.

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(J.L.A.-L.)

Calcutta

Calcutta (Bengali Kalikātā), from 1772 to 1912 the capital of British India and since 1947 the capital of West Bengal, is India's largest city and one of its major ports. The city is located on the east bank of the Hooghly River, once the main channel of the Ganges River, about 96 miles (154 kilometres) upstream from the head of the Bay of Bengal; there the port city developed as a point of transshipment from water to land and from river to sea.

The city proper has an area of about 40 square miles (104 square kilometres); however, the metropolitan area (Calcutta Urban Agglomeration) is much larger, consisting of about 533 square miles. A city of commerce, transport, and manufacture, Calcutta is the dominant urban centre of eastern India.

The city's name is an Anglicized version of Kalikātā. According to some, Kalikātā is derived from the Bengali word Kālīkshetra, meaning "Ground of (the goddess) Kālī." Some say the city's name derives from the location of its original settlement on the bank of a canal (*khāl*). A third opinion traces it to the Bengali words for lime (*kali*) and burnt shell (*kata*), since the area was noted for the manufacture of shell-lime. Still another opinion is that the name is derived from the Bengali term *kilkilā* (meaning, "flat area"), which is mentioned in the old literature.

This article is divided into the following sections:

Physical and human geography

THE CHARACTER OF THE CITY

Fashioned by the colonial British in the manner of a grand European capital—yet now set in one of the poorest and most overpopulated regions of India—Calcutta has grown into a city of sharp contrasts and contradictions. Calcutta has had to assimilate strong European influences and overcome the limitations of its colonial legacy in order to find its own unique identity. In the process it created an amalgam of East and West that found its expression in the life and works of the 19th-century Bengali elite and its most noteworthy figure, the poet and mystic Rabindranath Tagore.

This largest and most vibrant of Indian cities thrives amidst seemingly insurmountable economic, social, and political problems. Its citizens exhibit a great joie de vivre that is demonstrated in a penchant for art and culture and a level of intellectual vitality and political awareness unsurpassed in the rest of the country. No other Indian city can draw the kinds of crowds that throng to Calcutta's book fairs, art exhibitions, and concerts. There is a lively trading of polemics on walls, which has led to Calcutta being dubbed the "city of posters."

Yet for all of Calcutta's vitality, many of the city's residents live in some of the worst conditions, far removed from the cultural milieu. The city's energy, however, penetrates even to the meanest of slums, as a large number of Calcuttans sincerely support the efforts of those who minister to the poor and suffering. In short, Calcutta remains an enigma to many Indians as well as to foreigners. It continues to puzzle newcomers and to arouse an abiding nostalgia in the minds of those who have lived there.

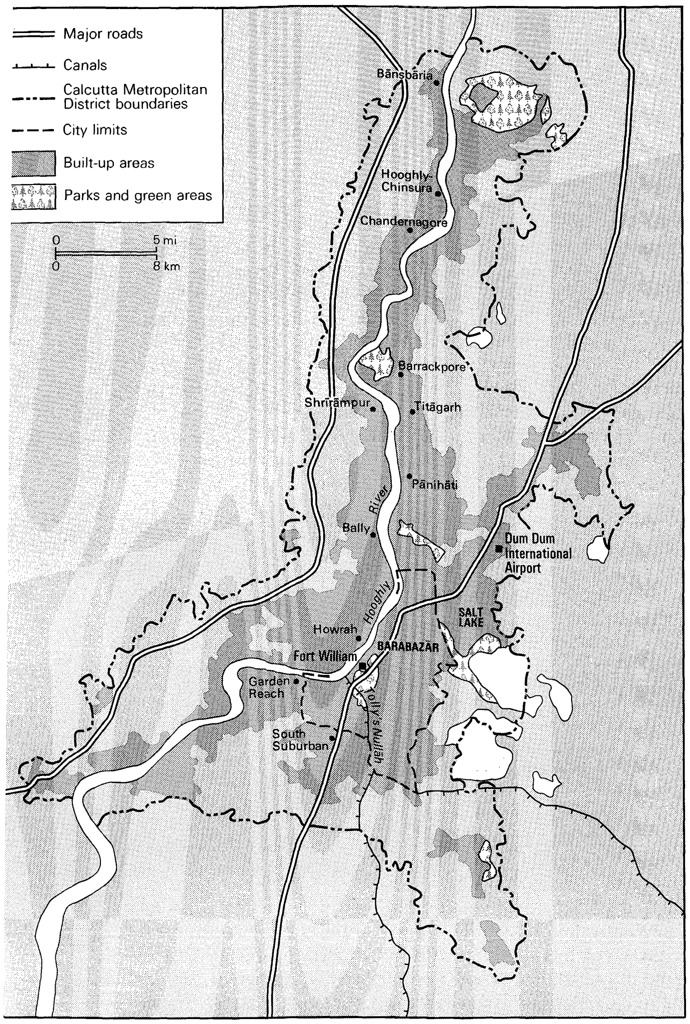
THE LANDSCAPE

The city site. The location of the city appears to have been originally selected partly because of its easily defensible position and partly because of its favourable trading location. The low, swampy, hot, and humid riverbank otherwise has little to recommend it. Its maximum elevation is about 30 feet (nine metres) above sea level. Eastward from the river the land slopes away to marshes and swamplands. Similar topography on the west bank of the river has confined the metropolitan area to a strip three to five miles wide on either bank of the river. Reclamation of the Salt Lake area on the northeastern fringe of the city demonstrated that the spatial expansion of the city is feasible, and further reclamation projects have been undertaken to the east, south, and west of the central area.

The principal suburbs of Calcutta are Howrah (on the west bank), Baranagar to the north, South Dum Dum to the northeast, the South Suburban Municipality (Behāla) to the south, and Garden Reach in the southwest. The whole urban complex is held together by close socioeconomic ties.

Climate. Calcutta has a subtropical climate with a seasonal regime of monsoons (rain-bearing winds). The maximum temperature reaches about 108° F (42° C) and the minimum temperature about 44° F (7° C). The average annual rainfall is about 64 inches (1,625 millimetres). Most of this falls from June to September, the period of the monsoon. These months are very humid and sometimes sultry. During October and November the rainfall dwindles. The winter months, from about the end of November to the end of February, are pleasant and rainless; fogs and mists occasionally reduce visibility in the early morning hours at this season, as also do thick blankets of smog in the evenings. The atmospheric pollution has greatly increased since the early 1950s. Factories, motor vehicles, and thermal-generating stations, which burn coal, are primary causes of this pollution, but monsoon winds act as cleansing agents by bringing in fresh air masses and also hastening the removal of water pollution.

The city layout. The most striking aspect of the layout of Calcutta is its rectangular, north–south orientation. With the exception of the central areas where Europeans formerly lived, the city has grown haphazardly. This haphazard development is most noticeable in the fringe areas around the central core formed by the city of Calcutta and the suburb of Howrah. The bulk of the city's administrative and commercial activity is concentrated in the Barabazār district, a small area north of the Maidān



Calcutta and surrounding area.

(the park containing Fort William and many of the city's cultural and recreational facilities). This has encouraged the development of a pattern of daily commuting that has overburdened Calcutta's transportation system, utilities, and other municipal facilities.

The street plan

Calcutta's system of streets and roads reflects the city's historical development. Local streets are narrow. There is only one express highway—Kāzi Nazrul Islām Avenue, which stretches from Calcutta to Dum Dum. The main roads form a grid pattern primarily in the old European sector, but elsewhere road planning has a haphazard character. Part of the reason for this has been the difficulty of providing enough river crossings; and it is for the same reason that most streets and highways run from north to south. Nullahs (watercourses) and canals that require bridging have also been important factors in influencing the road pattern.

The housing shortage

Housing. The city has an acute housing shortage. Of the persons living in institutional shelters in the Calcutta Metropolitan District, more than two-thirds live in the city itself. About three-fourths of the housing units in the city are used for dwelling purposes only. There are hundreds of bustees, or slums, where about one-third of the city's population lives. A bustee is officially defined as "a collection of huts standing on a plot of land of at least one-sixth of an acre." There also are bustees built on less than one-sixth of an acre (one-fifteenth of a hectare). The majority of huts are tiny, unventilated, single-story rooms, often dilapidated. They have few sanitary facilities, and there is very little open space. The government sponsors a bustee-improvement program.

Architecture. In contemporary Calcutta the skyline is broken in some areas by skyscrapers and tall multistory blocks. The cityscape has changed rapidly. The Chowringhee area in central Calcutta, once a row of palatial houses, has been given up to offices, hotels, and shops. In northern and central Calcutta, buildings are still mainly two or three stories high. In southern and south central Calcutta, multistoried apartment buildings have become more common.

Western architectural influences

Western influence is dominant in Calcutta's architectural monuments. The Rāj Bhavan (the state governor's residence) is an imitation of Kedleston Hall in Derbyshire; the High Court resembles the Cloth Hall at Ypres, Belg.; the Town Hall is in Grecian style with a Doric-Hellenic portico; St. Paul's Cathedral is of Indo-Gothic-style architecture; the Writers' Building is of Gothic-style architecture with statuary on top; the Indian Museum is in an Italian style; and the General Post Office, with its majestic dome, has Corinthian columns. The beautiful column of the Sahid Minār (Ochterlony Monument) is 165 feet high—its base is Egyptian, its column Syrian, and its cupola in the Turkish style. The Victoria Memorial represents an attempt to combine classical Western influence with Mughal architecture; the Nakhoda Mosque is modeled on Akbar's Tomb at Sikandra; the Birla Planetarium is based on the stupa (Buddhist reliquary) at Sānchī.

The West Bengal Legislative Council is a dignified building in the modern architectural style. The Ramakrishna Mission Institute of Culture, the most important example of postindependence construction, follows the style of ancient Hindu palace architecture in northwestern India.

THE PEOPLE

The density of population is extremely high, about 85,500 persons per square mile (33,000 per square kilometre). Overcrowding has reached virtually intolerable proportions in many sections of the city. Calcutta has experienced a high rate of population growth for more than a century, but events such as the partitioning of Bengal in 1947 and warfare in Bangladesh in the early 1970s precipitated massive population influxes. Large refugee colonies have also sprung up in the northern and southern suburbs. In addition, a large number of migrants from other states—mostly from neighbouring Bihār and Orissa and eastern Uttar Pradesh—have come to Calcutta in search of employment.

More than four-fifths of the population is Hindu. Muslims and Christians constitute the largest minorities, but there are some Sikhs, Jains, and Buddhists. The dominant language is Bengali, but Urdu, Oriya, Tamil, Punjabi, and other languages are also spoken. Calcutta is also a cosmopolitan city: other groups present include a variety of peoples from Asia (notably Bangladeshis and Chinese), Europeans, North Americans, and Australians. Calcutta was racially segregated under British rule, the Europeans living in the city centre and Indians living to the north and south. The pattern of segregation has continued in the modern city, although the distribution is now based on religious, linguistic, educational, and economic criteria. Slums and low-income residential areas, however, exist side-by-side with more affluent areas.

THE ECONOMY

Calcutta's position as one of India's preeminent economic centres is rooted in its industries, financial and trade activities, and role as a major port; it is also a major centre for printing, publishing, and newspaper circulation, as well as for recreation and entertainment. The products of Calcutta's hinterland include coal, iron, manganese, mica, petroleum, tea, and jute. Unemployment has been a continuing and growing problem since the 1950s. Unemployment in Calcutta is, to a large extent, a problem of the college-educated and of people trained for clerical and other white-collar jobs.

Manufactured products

Industry. Calcutta is the world's largest processor of jute. The jute industry was established in the 1870s, and mills now extend north and south of the city centre on both banks of the Hooghly River. Engineering constitutes the city's other major industry. In addition, city factories produce and distribute a variety of consumer goods—notably foodstuffs, beverages, tobacco, and textiles—other light manufactures, and chemicals. Calcutta's industries have been in a general decline since Indian independence in 1947. Major factors contributing to this decline have been the loss of the eastern part of Bengal at independence, an overall decline in Calcutta's industrial productivity, and the lack of industrial diversification in the city.

Finance and trade. The Calcutta stock exchange plays an important part in the organized financial market of the country. Foreign banks also have a significant business base in Calcutta although the city's importance as an international banking centre has declined. In addition, the controlling agency for coal mines is in Calcutta. Jute mills and large-scale engineering industries are also controlled from the city; and the Bengal Chamber of Commerce and Industry, the Bengal National Chamber of Commerce and Industry, and the Indian Chamber of Commerce are based there.

That the city's economy is primarily mercantile is reflected in the fact that about two-fifths of the workers are employed in trade and commerce. Other important occupations include public-sector service in government departments, financial institutions, and medical and educational institutions. Private-sector services include the stock exchange, medical and educational services, legal services, accountancy and credit firms, chambers of commerce, and various utility services.

Transportation. The condition of the surfaced roads in the city is poor, although the traffic load is heavy. The mass-transportation system mainly depends on trams and buses. Trams are under government management. Buses are run by the government and by private companies. In 1986 the first section of a subway system—the first in India—was opened in the city.

The connection between Calcutta and its hinterland to the west depends upon only a few bridges over the Hooghly—the Howrah Bridge and, farther north, the bridges at Bally and Naihāti. The Howrah Bridge, Calcutta's main link with the hinterland, carries eight lanes of vehicular traffic, has two tramway tracks in the centre, and is one of the most heavily used bridges in the world. A second bridge between Howrah and Calcutta has been under construction since the 1970s.

The Grand Trunk Road (National Highway No. 2) is one of the oldest road routes in India. It runs from Howrah to Kashmir and is the main route connecting the city with northern India. Other national highways connect Calcutta with the west coast of India, the northern part of West Bengal, and the frontier with Bangladesh.

Two railway terminals—Howrah on the west bank and Sealdah on the east—serve the railway networks running north and south as well as those running east and west. Calcutta's major air terminal, at Dum Dum, handles international and domestic flights.

The Calcutta port handles—in terms of volume—one-tenth of India's import cargoes and about one-twelfth of its export cargoes. Some decline in traffic has occurred, however, partly because of problems encountered in dredging silt from the river and partly because of labour problems. Transport, storage, wholesaling, and retailing requirements for exports and imports are concentrated in Calcutta and Howrah. The Calcutta port lost its position as India's preeminent cargo handler in the 1960s, but it and the port of Haldia (about 40 miles downstream) still account for a large portion of the country's foreign exchange.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Government in the city proper is the responsibility of the Calcutta Municipal Corporation; the corporation's council is composed of one elected representative from each of the city's 100 wards. The council members annually elect a mayor, a deputy mayor, and a number of committees to conduct the activities of the corporation. A commissioner, the executive head of the corporation, is responsible to its elected membership. The city is also a part of the Calcutta Metropolitan District, an entity created to oversee planning and development on a regional basis. This district includes a large rural hinterland around the urban centres.

Because Calcutta is the capital of West Bengal, the governor resides in the city in the historic Rāj Bhavan. The state Legislative Assembly is located in the city, as is the Secretariat, housed in the Writers' Building, with the state ministries in charge of various departments. The Calcutta High Court, exercising original jurisdiction over the city and appellate jurisdiction over West Bengal, is also located there. A number of national government institutions—including the National Library, the Indian Museum, and the Geological Survey of India—are also in the city.

Services. Filtered water is supplied from the main waterworks located outside the city at Palta, as well as from some 200 major wells and 3,000 smaller ones. The Farakka Barrage (dam) on the Ganges, 240 miles upriver from Calcutta, ensures a generally saline-free water supply for the city; but because existing water supplies are inadequate, salinity continues to be a problem during the dry months. In addition, unfiltered water, supplied daily for watering the city streets and for the fire brigade, is used by many residents for their daily needs. This circumstance was largely responsible for the former prevalence of cholera during the summer months, but chlorination of unfiltered water and cholera inoculation have reduced considerably the occurrence of the disease.

Water supplies

Municipal Calcutta has several hundred miles of sewers and surface drains, but much of the city remains unsupplied with sewers. Accumulation of silt has narrowed many sewer channels. Unsanitary methods of human-waste disposal persist in the unsewered areas of the city. The system of removing garbage and of garbage dumping is also unsatisfactory.

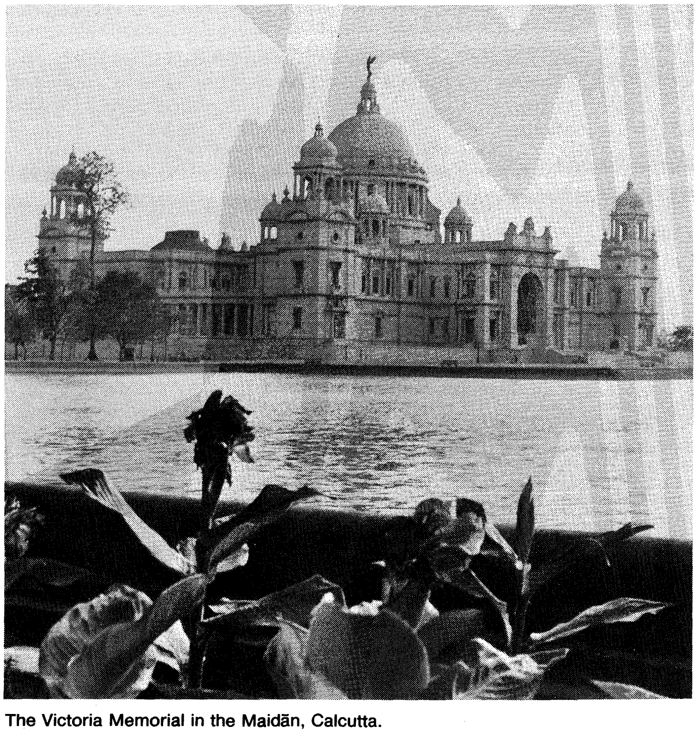
Calcutta is supplied with electricity by a variety of sources, including the Calcutta Electric Supply Corporation; West Bengal State Electricity Board; Durgapur Projects, Ltd.; Bandel Thermal Power Station; Santaldih Power Station; and Damodar Valley Corporation Grid. There is still a gap, however, between generating capacity and potential demand, and temporary power interruptions occur on occasion.

Security

Administration of the Calcutta police force is vested in the city's commissioner of police, as is direction of the suburban police force. The city is divided into four police divisions. The fire brigade has its headquarters in central Calcutta.

Health. Smallpox has been completely eradicated from the city, and death from malaria and enteric fever has been brought under control. The incidence of tuberculosis has also declined. Hundreds of hospitals, private clinics, free dispensaries run by the Calcutta Municipal Corpora-

G.R. Richardson—Taurus Photos



The Victoria Memorial in the Maidān, Calcutta.

tion and charitable trusts, and state-operated polyclinics serve the Calcutta region. The Order of the Missionaries of Charity, an organization founded (1948) by Mother Teresa (recipient of the Nobel Peace Prize of 1979), cares for the blind, the aged, the dying, and lepers in the poorest sections of the city. There are several medical colleges in addition to other medical research centres. The number of doctors per 1,000 persons is greater in Calcutta than in most parts of the country, but their distribution is uneven; and since the city is a medical centre for the northeastern region of India, its health-care facilities are always overcrowded.

Medical acilities

Education. Education has long been a mark of higher social status in Calcutta. The city has been a centre of learning since the resurgence in Indian education that began in Bengal in the early 19th century. The first English-style school, the Hindu College (later called Presidency College), was founded in 1817.

Primary education is supervised by the government of West Bengal and is free in schools run by the municipal corporation. A large number of children, however, attend recognized schools that are under private management. Most secondary schools are under the supervision of the state, but some are accredited through the national government and a few through the British educational system.

The universities

Calcutta has three major universities: the University of Calcutta, Jadavpur University, and Rabindra Bharati University. The University of Calcutta, founded in 1857, has more than 150 affiliated colleges. Besides these colleges, university colleges of arts (humanities), commerce, law, medicine, science, and technology specialize in postgraduate teaching and research. Jadavpur University has three faculties—arts (humanities), science, and engineering. Although the university has a small number of colleges affiliated with it, its main focus is on graduate and post-graduate instruction on a single campus. Rabindra Bharati University specializes in humanities and the fine arts (dance, drama, and music).

Research institutions include the Indian Statistical Institute, the Indian Association for the Cultivation of Science, the Bose Institute (natural science), and the All-India Institute of Hygiene and Public Health, which is a constituent college of the University of Calcutta.

CULTURAL LIFE

Calcutta is the most important cultural centre of India. The city is the birthplace of modern Indian literary and artistic thought and of Indian nationalism, and the efforts of its citizens to preserve Indian culture and civilization have no parallel in the rest of the country. The blending of Eastern and Western cultural influences over the centuries has stimulated the creation of numerous and diverse organizations that contribute to Calcutta's cultural life. In addition to the three universities, these include the Asiatic Society, the Bengal Literary Society (Bangiya Sahitya Parishad), the Ramakrishna Mission Institute of Culture, the Academy of Fine Arts, the Birla Academy of Art and Culture, and the Maha Bodhi Society.

Museums and libraries. Greater Calcutta has more than 30 museums, which cover a wide variety of fields. The Indian Museum, founded in 1814, is the oldest in India and is the largest museum of its kind in the country; the archaeology and numismatic sections contain the most valuable collections. The exhibits at Victoria Memorial trace Britain's relations with India. The Asutosh Museum of Indian Art in the University of Calcutta has exhibits of the folk art of Bengal among its collections. Valuable library collections are to be found in the Asiatic Society, Bengal Literary Society, and the University of Calcutta; the National Library is the largest in India and contains a fine collection of rare books and manuscripts.

The arts. Calcuttans have long been active in literary and artistic pursuits. The literary movement spawned there in the mid-19th century through exposure to Western forms sparked a cultural renaissance throughout India. The best exponent of this movement was Rabindranath Tagore, winner of the Nobel Prize for Literature of 1913, whose remarkable creativity in poetry, music, drama, and painting continues to enrich the cultural life of the city. Calcutta remains at the vanguard of artistic movements in the country, and several artists' societies present annual shows.

Calcutta is also a centre of traditional and contemporary music and dance. In 1937 Tagore inaugurated the first All-Bengal Music Conference in Calcutta. Since then, a number of classical Indian music conferences have been held every year. The home of many classical dancers, Calcutta was also the location of Uday Shankar's experiments at adapting Western theatrical techniques to traditional dance forms. The school of dance, music, and drama founded by him has been in the city since 1965.

Theatre and films

Professional drama got its start in Calcutta in the 1870s with the founding of the National Theatre. Modern dramatic forms were pioneered in the city by such playwrights as Girish Chandra Ghosh and Dirabandhu Mitra. Calcutta is still an important centre of professional and amateur theatre and of experimental drama. The city has also been a pioneering centre of motion-picture production in India. The avant-garde film directors Satyajit Ray and Mrinal Sen have achieved international acclaim. There are scores of cinemas in the city, which regularly show films in English, Bengali, and Hindi.

Recreation. More than 200 parks, squares, and open spaces are maintained by the Calcutta Municipal Corporation. There is, however, very little open space in the overcrowded parts of the city. The Maidān, about two miles in length and a mile in width, is the best-known open space; the major soccer, cricket, and hockey fields are located there. Adjacent to the Maidān is one of the oldest cricket fields in the world, Ranji Stadium, in the Eden Gardens; Netāji Stadium, for indoor events, is also in the vicinity. The Salt Lake Stadium, built to the east of the city, can seat 100,000 spectators. There are two racecourses and two golf courses within the city, and rowing at the Lake Club and the Bengal Rowing Club is popular. The Zoological Gardens are spread over an area of some 50 acres (20 hectares). The Indian Botanical Gardens is located on the west bank of the rivers; the herbarium there contains about 40,000 species of plants.

History

THE EARLY PERIOD

The name Kalikātā had been mentioned in the rent-roll of the Mughal emperor Akbar (reigned 1556–1605) and also in the *Manasā-maṅgal* of the Bengali poet Bipradās (1495). The history of Calcutta as a British settlement dates from the establishment of a trading post there by Job Charnock, an agent of the English East India Company, in 1690. Charnock had previously had disputes with officials of the Mughal Empire at the river port of Hooghly and had been obliged to leave, after which he attempted unsuccessfully to establish himself at other places down the river. When the Mughal officials, not wishing to lose what they had gained from the English company's commerce, permitted Charnock to return once more, he chose Calcutta as the seat of his operations. The site was apparently carefully selected, being protected by the Hooghly River on the west, a creek to the north, and by salt lakes about two and a half miles to the east. Rival Dutch, French, and other European settlements were higher up the river on the west bank, so that access from the sea was not threatened, as it was at the port of Hooghly. The river at this point was also wide and deep; the only disadvantage was that the marshes to the east and swamps within the area made the spot unhealthy. Moreover, before the coming of the English, three local villages—Sutānati, Kalikātā, and Gobindapore, which were later to become parts of Calcutta—had been chosen as places to settle by Indian merchants who had migrated from the silted-up port of Satgāon, farther upstream. The presence of these merchants may have been to some extent responsible for Charnock's choice of the site.

By 1696, when a rebellion broke out in the nearby district of Burdwān, the Mughal provincial administration had become friendly to the growing settlement. The servants of the company, who asked for permission to fortify their trading post, or factory, were given permission in general terms to defend themselves. The rebels were easily crushed by the Mughal government, but the settlers' defensive structure of brick and mud remained and, in 1700, came to be known as Fort William. In 1698 the English obtained letters patent that granted them the privilege of purchasing the zamindari right (the right of revenue collection; in effect, the ownership) of the three villages.

GROWTH OF THE CITY

In 1717 the Mughal emperor Farrukh-siyar granted the East India Company freedom of trade in return for a yearly payment of 3,000 rupees; this arrangement gave a great impetus to the growth of Calcutta. A large number of Indian merchants flocked to the city. The servants of the company, under the company's flag, carried on a duty-free private trade. When the Marāṭhās from the southwest began incursions against the Mughals in the western districts of Bengal in 1742, the English obtained permission from ʿAli Vardi Khan, the nawab (ruler) of Bengal, to dig an entrenchment in the northern and eastern part of the town to form a moat on the land side. This came to be known as the Marāṭhā Ditch. Although it was not completed to the southern end of the settlement, it marked the city's eastern boundary.

In 1756 the Nawab's successor, Sirāj-ud-Dawlah, captured the fort and sacked the town. Calcutta was recaptured in January 1757 by Robert Clive, one of the founders of British power in India, and by the British admiral Charles Watson. The Nawab was defeated shortly afterward at Plassey (June 1757), after which British rule in Bengal was assured. Gobindapore was cleared of its forests, and the new Fort William was built on its present site, overlooking the Hooghly at Calcutta, where it became the symbol of British military ascendancy.

The capital of British India

Calcutta did not become the capital of British India until 1772, when the first governor general, Warren Hastings, transferred all important offices to the city from Murshidābād, the provincial Mughal capital. In 1773 Bombay and Madras became subordinate to the government at Fort William. A supreme court administering English law began to exercise original jurisdiction over the city as far as the Marāṭhā Ditch (now Ācharya Prafulla Chandra and Jagadish Chandra Bose roads).

In 1706 the population of Calcutta had been about 10,000 to 12,000. By 1752 it had increased to nearly 120,000 and by 1821 to 180,000. The White (British) Town was built on ground that had been raised and drained. There were so many palaces in the British sector of the city that it was named the "city of palaces." Outside the British town were built the mansions of the newly rich, as well as clusters of huts. The names of different quarters of the city—such as Kumārtuli (the potters' district) and Sankaripara (the conch-shell workers' district)—still indicate the various occupational castes of the people who became residents of the growing metropolis. Two distinct areas—one British, one Indian—came to coexist in Calcutta.

Calcutta at this time was described as a pestilential town. There were few good roads. In 1814 a Lottery Committee was constituted to finance public improvement by means of lotteries, and between 1814 and 1836 it took some effective measures to improve conditions. The corporation was established in 1841. Cyclones in 1864, 1867, and 1870, however, devastated the poorer, low-lying areas.

Development of the port

By successive stages, as British power extended over the subcontinent, the whole of northern India became a hinterland for the port of Calcutta. The abolition of inland customs duties in 1835 created an open market, and the construction of railways (beginning in 1854) further quickened the development of business and industry. It was at this time that the Grand Trunk Road from Calcutta to Peshāwar (now in Pakistan) was completed. British mercantile, banking, and insurance interests flourished. The Indian sector of Calcutta also became a busy hub of commerce and was thronged with people from all parts of India and many other parts of Asia. Calcutta became the intellectual centre of the subcontinent.

CALCUTTA IN THE 20TH CENTURY

The 20th century marked the beginning of Calcutta's woes. Lord Curzon, viceroy of India, partitioned Bengal in 1905, making Dacca (now Dhākā) the capital of eastern Bengal and Assam. Insistent agitation led to the annulment of this partition, but in 1912 the capital of British India was removed from Calcutta to Delhi, where the government could enjoy relative calm. The partition of Bengal in 1947 was a final blow.

Modern social problems

As Calcutta's population grew larger, social problems also became more insistent, as did demands for home rule for India. Communal riots occurred in 1926, and, when Mahatma Gandhi called for noncompliance with unjust laws, riots occurred in 1930. In World War II, Japanese air raids upon the Calcutta docks caused damage and loss of life. The most serious communal riots of all took place in 1946, when the partition of British India became imminent and tensions between Muslims and Hindus reached their height.

In 1947 the partition of Bengal between India and Pakistan constituted a serious setback for Calcutta, which became the capital of West Bengal only, losing the trade of a part of its former hinterland. At the same time, millions of refugees from East Pakistan (now Bangladesh) flocked to Calcutta, aggravating social problems and increasing overcrowding, which had already assumed serious proportions. Economic stagnation in the mid-1960s further increased the instability of the city's social and political life and fueled a flight of capital from the city. The management of many companies was assumed by the state government. Particularly in the 1980s, large-scale public works programs and centralized regional planning contributed to the improvement of economic and social conditions in the city.

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(N.K.S./S.B.-G.)

Calendar

A calendar is a means of grouping days in ways convenient for regulating civil life and religious observances and for historical and scientific purposes. The word is derived from the Latin *calendarium,* meaning interest register, or account book, itself a derivation from *calendae* (or *kalendae*), the first day of the Roman month, the day on which future market days, feasts, and other occasions were proclaimed.

The development of a calendar is vital for the study of chronology, since this is concerned with reckoning time by regular divisions, or periods, and using these to date events. It is essential, too, for any civilization that needs to measure periods for agricultural, business, domestic, or other reasons. The first practical calendar to evolve from these requirements was the Egyptian, and it was this that the Romans developed into the Julian calendar that served western Europe for more than 1,500 years. The Gregorian calendar was a further improvement and has been almost universally adopted because it satisfactorily draws into one system the dating of religious festivals based on the phases of the Moon and seasonal activities determined by the movement of the Sun. Such a calendar system is complex, since the periods of the Moon's phases and the Sun's motion are incompatible; but by adopting regular cycles of days and comparatively simple rules for their application, the calendar provides a year with an error of less than half a minute.

This article is divided into the following sections:

Measurement of time and types of calendars

STANDARD UNITS AND CYCLES

The basic unit of computation in a calendar is the day, and although days are now measured from midnight to midnight this has not always been so. Astronomers, for instance, from about the 2nd century AD until 1925 counted days from noon to noon. In earlier civilizations and among primitive peoples, where there was less communication between different settlements or groups, different methods of reckoning the day presented no difficulties. Most primitive tribes used a dawn-to-dawn reckoning, calling a succession of days so many dawns, or suns. Later, the Babylonians, Jews, and Greeks counted a day from sunset to sunset, whereas the day was said to begin at dawn for the Hindus and Egyptians and at midnight for the Romans. The Teutons counted nights, and from them the grouping of 14 days called a fortnight is derived.

Divisions of the day

There was also great variety in the ways in which the day was subdivided. In Babylonia, for example, the astronomical day was divided differently than the civil day, which, as in other ancient cultures, was composed of "watches." The length of the watches was not constant but varied with the season, the day watches being the longer in summer and the night watches in the winter. Such seasonal rariations in divisions of the day, now called seasonal or temporal hours, became customary in antiquity because they corresponded to the length of the Sun's time above the horizon, at maximum in summer and at minimum in winter. Only with the advent of mechanical clocks in western Europe at the end of the 13th century did seasonal (unequal) hours become inconvenient.

Most early Western civilizations used 24 seasonal hours in the day—12 hours of daylight and 12 of darkness. This was the practice of the Greeks, the Sumerians and Babylonians, the Egyptians, and the Romans, and of Western Christendom so far as civil reckoning was concerned. The church adopted its own canonical hours for reckoning daily worship: there were seven of these—matins, prime, terce, sext, none, vespers, and compline—but in secular affairs the system of 24 hours held sway. This number, 2 × 12, or 24, was derived in Babylonia from the Sumerian sexagesimal method of reckoning, based on gradations of 60 ($5 \times 12 = 60$) rather than on multiples of 10. In Babylonia, for most purposes, both daylight and night were divided into three equal watches, and each watch was subdivided into half- and quarter-watches. Babylonian astronomers, perhaps in preference to the variable civil system, divided every day into 12 equal units, called *bēru,* each of which was subdivided into 30 *gesh.* The oldest known astronomical texts are from the Old Babylonian period, but this dual system may be attributable to earlier Sumerian society.

Origins of the week

Once the day is divided into parts, the next task is to gather numbers of days into groups. Among primitive peoples, it was common to count moons (months) rather than days, but later a period shorter than the month was thought more convenient, and an interval between market days was adopted. In West Africa some tribes used a four-day interval; in central Asia five days was customary; the Assyrians adopted five days and the Egyptians, 10 days, whereas the Babylonians attached significance to the days of the lunation that were multiples of seven. In ancient Rome, markets were held at eight-day intervals; because of the Roman method of inclusive numeration, the market day was denoted *nundinae* ("ninth-day") and the eight-day week, an *inter nundium.*

The seven-day week may owe its origin partly to the four (approximately) seven-day phases of the Moon and partly to the Babylonian belief in the sacredness of the number seven, which was probably related to the seven planets. Moreover, by the 1st century BC the Jewish seven-day week seems to have been adopted throughout the Roman world, and this influenced Christendom. The names in English of the days of the week are derived from Latin or Anglo-Saxon names of gods.

The month is based on the lunation, that period in which the Moon completes a cycle of its phases. The period lasts approximately 29½ days, and it is easy to recognize and short enough for the days to be counted without using large numbers. In addition, it is very close to the average menstrual period of women and also to the duration of cyclic behaviour in some marine creatures. Thus, the month possessed great significance and was often the governing period for religious observances, of which the dating of Easter is a notable example. Most early calendars were, essentially, collections of months, the Babylonians using 29- and 30-day periods alternately, the Egyptians fixing the duration of all months at 30 days, with the Greeks copying them, and the Romans in the Julian calendar having a rather more complex system using one 28-day period with the others of either 30 or 31 days.

The seasons

The month is not suitable for determining the seasons, for these are a solar, not a lunar, phenomenon. Seasons vary in different parts of the world—in tropical countries there are just the rainy and dry periods, but elsewhere there are successions of wider changes. In Egypt the annual flooding of the Nile was followed by seeding and then harvest, and three seasons were recognized; but in Greece and other more northern countries there was a succession of four seasons of slightly different lengths. However many there seemed to be, it was everywhere recognized that seasons were related to the Sun and that they could be determined from solar observations. These might consist of noting the varying length of the midday shadow cast by a stick thrust vertically into the ground or follow the far more sophisticated procedure of deducing from nocturnal observations the Sun's position against the background of the stars. In either case the result was a year of 365 days, a period incompatible with the 29½-day lunation. To find some simple relationship between the two periods was the problem that faced all calendar makers from Babylonian times onward.

Nonastronomical cycles

A number of nonastronomical natural signs have also been used in determining the seasons. In the Mediterranean area, such indications change rapidly, and Hesiod (*c.* 800 BC) mentions a wide variety: the cry of migrating cranes, which indicated a time for plowing and sowing; the time when snails climb up plants, after which digging in vineyards should cease; and so on. An unwitting approximation to the tropical year may also be obtained by intercalation, using a simple lunar calendar and observations of animal behaviour. Such an unusual situation has grown up among the Yami fishermen of Botel-Tobago Island, near Taiwan. They use a calendar based on phases of the Moon, and some time about March (the precise date depends on the degree of error of their lunar calendar compared with the tropical year) they go out in boats with lighted flares. If flying fish appear, the fishing season is allowed to commence, but if the lunar calendar is too far out of step with the seasons, the flying fish will not rise. Fishing is then postponed for another lunation, which they insert in the lunar calendar, thus having a year of 13 instead of the usual 12 lunations.

TIME DETERMINATION BY STARS, SUN, AND MOON

Measurement of the day, the month, and the year

Celestial bodies provide the basic standards for determining the periods of a calendar. Their movement as they rise and set is now known to be a reflection of the Earth's rotation, which, although not precisely uniform, can conveniently be averaged out to provide a suitable calendar day. The day can be measured either by the stars or by the Sun. If the stars are used, then the interval is called the sidereal day and is defined by the period between two passages of a star (more precisely of the vernal equinox, a reference point on the celestial sphere) across the meridian: it is 23 hours 56 minutes 4.10 seconds of mean solar time (see below). The interval between two passages of the Sun across the meridian is a solar day. In practice, since the rate of the Sun's motion varies with the seasons, use is made of a fictitious Sun that always moves across the sky at an even rate. This period of constant length, far more convenient for civil purposes, is the mean solar day, which has a duration in sidereal time of 24 hours 3 minutes 56.55 seconds. It is longer than the sidereal day because the motion of the Earth in its orbit during the period between two transits of the Sun means that the Earth must complete more than a whole revolution to bring the Sun back to the meridian. The mean solar day is the period used in calendar computation.

The month is determined by the Moon's passage around the Earth, and, as in the case of the day, there are several ways in which it can be defined. In essence, these are of two kinds: first, the period taken by the Moon to complete an orbit of the Earth and, second, the time taken by the Moon to complete a cycle of phases. Among primitive societies, the month was determined from the phases; this interval, the synodic month, is now known to be 29.53059 days. The synodic month grew to be the basis of the calendar month.

The year is the period taken by the Earth to complete an orbit around the Sun and, again, there are a number of ways in which this can be measured. But for calculating a calendar that is to remain in step with the seasons, it is most convenient to use the tropical year, since this refers directly to the Sun's apparent annual motion. The tropical year is defined as the interval between successive passages of the Sun through the vernal equinox (*i.e.,* when it crosses the celestial equator late in March) and amounts to 365.242199 mean solar days.

The tropical year and the synodic month are incommensurable, 12 synodic months amounting to 354.36706 days, almost 11 days shorter than the tropical year. Moreover, neither is composed of a complete number of days, so that to compile any calendar that keeps in step with the Moon's phases or with the seasons it is necessary to insert days at appropriate intervals; such additions are known as intercalations.

Locating the Sun among the stars

In primitive lunar calendars, intercalation was often achieved by taking alternately months of 29 and 30 days. When, in order to keep dates in step with the seasons, a solar calendar was adopted, some greater difference between the months and the Moon's phases was bound to occur. And the solar calendar presented an even more fundamental problem—that of finding the precise length of the tropical year. Observations of cyclic changes in plant or animal life were far too inaccurate, and astronomical observations became necessary. Since the stars are not visible when the Sun is in the sky, some indirect way had to be found to determine its precise location among them. In tropical and subtropical countries it was possible to use the method of heliacal risings. Here the first task was to determine the constellations around the whole sky through which the Sun appears to move in the course of a year. Then, by observing the stars rising in the east just after sunset it was possible to know which were precisely opposite in the sky, where the Sun lay at that time. Such heliacal risings could, therefore, be used to determine the seasons and the tropical year. In temperate countries, the angle at which stars rise up from the horizon is not steep enough for this method to be adopted, so that there wood or stone structures were built to mark out points along the horizon to allow analogous observations to be made.

Stonehenge

The most famous of these is Stonehenge in Wiltshire, England, where the original structure appears to have been built about 2000 BC and additions made at intervals several centuries later. It is composed of a series of holes, stones, and archways arranged mostly in circles, the outermost ring of holes having 56 marked positions, the inner ones 30 and 29, respectively. In addition, there is a large stone—the heel stone—set to the northeast, as well as some smaller stone markers. Observations were made by lining up holes or stones with the heel stone or one of the other markers and watching for the appearance of the Sun or Moon against that point on the horizon that lay in the same straight line. The extreme north and south positions on the horizon of the Sun—the summer and winter solstices—were particularly noted, while the inner circles, with their 29 and 30 marked positions, allowed "hollow" and "full" (29- or 30-day) lunar months to be counted off. More than 600 contemporaneous structures of an analogous but simpler kind have been discovered in Britain, in Brittany, and elsewhere in Europe and the Americas. It appears, then, that astronomical observation for calendrical purposes was a widespread practice in some temperate countries three to four millennia ago.

Today a solar calendar is kept in step with the seasons by a fixed rule of intercalation. But although the Egyptians, who used the heliacal rising of Sirius to determine the annual inundation of the Nile, knew that the tropical year was about 365.25 days in length, they still used a 365-day year without intercalation. This meant that the calendar date of Sirius' rising became increasingly out of step with the original dates as the years progressed. In consequence, while the agricultural seasons were regulated by the heliacal rising of Sirius, the civil calendar ran its own separate course. It was not until well into Roman times that an intercalary day once every four years was instituted to retain coincidence.

COMPLEX CYCLES

The fact that neither months nor years occupied a whole number of days was recognized quite early in all the great civilizations. Some observers also realized that the difference between calendar dates and the celestial phenomena due to occur on them would first increase and then diminish until the two were once more in coincidence. The succession of differences and coincidences would be cyclic, recurring time and again as the years passed. An early recognition of this was the Egyptian Sothic cycle, based on the star Sirius (called Sothis by ancient Egyptians). The error with respect to the 365-day year and the heliacal risings of Sirius amounted to one day every four tropical years, or one whole Egyptian calendar year every 1,460 tropical years (4 × 365), which was equivalent to 1,461 Egyptian calendar years. After this period the heliacal rising and setting of Sothis would again coincide with the calendar dates (see below *The Egyptian calendar*).

The *octaëteris*

The main use of cycles was to try to find some commensurable basis for lunar and solar calendars, and the best known of all the early attempts was the *octaëteris*, usually attributed to Cleostratus of Tenedos (*c.* 500 BC) and Eudoxus of Cnidus (390–*c.* 340 BC). The cycle covered eight years, as its name implies, and so the *octaëteris* amounted to 8 × 365, or 2,920 days. This was very close to the total of 99 lunations (99 × 29.5 = 2,920.5 days), so this cycle gave a worthwhile link between lunar and solar calendars. When, in the 4th century BC, the accepted length of the year became 365.25 days, the total number of solar calendar days involved became 2,922, and it was then realized that the *octaëteris* was not as satisfactory a cycle as supposed.

The saros

Another early and important cycle was the saros, essentially an eclipse cycle. There has been some confusion over its precise nature because the name is derived from the Babylonian word *shār* or *shāru,* which could mean either "universe" or the number 3,600 (*i.e.,* 60 × 60). In the latter sense it was used by Berosus (*c.* 290 BC) and a few later authors to refer to a period of 3,600 years. What is now known as the saros and appears as such in astronomical textbooks (still usually credited to the Babylonians) is a period of 18 years $11\frac{1}{3}$ days (or with one day more or less, depending on how many leap years are involved), after which a series of eclipses is repeated.

The Metonic cycle

In Central America an independent system of cycles was established (see below *Ancient and religious calendar systems: The Americas*). The most significant of all the early attempts to provide some commensurability between a religious lunar calendar and the tropical year was the Metonic cycle. This was first devised about 432 BC by the astronomer Meton of Athens. Meton worked with another Athenian astronomer, Euctemon, and made a series of observations of the solstices, when the Sun's noonday shadow cast by a vertical pillar, or gnomon, reaches its annual maximum or minimum, to determine the length of the tropical year. Taking a synodic month to be 29.5 days, they then computed the difference between 12 of these lunations and their tropical year, which amounted to 11 days. It could be removed by intercalating a month of 33 days every third year. But Meton and Euctemon wanted a long-term rule that would be as accurate as they could make it, and they therefore settled on a 19-year cycle. This cycle consisted of 12 years of 12 lunar months each and seven years each of 13 lunar months, a total of 235 lunar months. If this total of 235 lunations is taken to contain 110 hollow months of 29 days and 125 full months of 30 days, the total comes to (110 × 29) + (125 × 30), or 6,940 days. The difference between this lunar calendar and a solar calendar of 365 days amounted to only five days in 19 years and, in addition, gave an average length for the tropical year of 365.25 days, a much-improved value that was, however, allowed to make no difference to daily reckoning in the civil calendar. But the greatest advantage of this cycle was that it laid down a lunar calendar that possessed a definite rule for inserting intercalary months and kept in step with a cycle of the tropical years. It also gave a more accurate average value for the tropical year and was so successful that it formed the basis of the calendar adopted in the Seleucid Empire (Mesopotamia) and was used in the Jewish calendar and the calendar of the Christian Church; it also influenced Indian astronomical teaching.

The Metonic cycle was improved by both Callippus and Hipparchus. Callippus of Cyzicus (*c.* 370–300 BC) was perhaps the foremost astronomer of his day. He formed what has been called the Callippic period, essentially a cycle of four Metonic periods. It was more accurate than the original Metonic cycle and made use of the fact that 365.25 days is a more precise value for the tropical year than 365 days. The Callippic period consisted of 4 × 235, or 940 lunar months, but its distribution of hollow and full months was different from Meton's. Instead of having totals of 440 hollow and 500 full months, Callippus adopted 441 hollow and 499 full, thus reducing the length of four Metonic cycles by one day. The total days involved therefore became (441 × 29) + (499 × 30), or 27,759, and 27,759 ÷ (19 × 4) gives 365.25 days exactly. Thus the Callippic cycle fitted 940 lunar months precisely to 76 tropical years of 365.25 days.

Precession of the equinoxes

Hipparchus, who flourished in Rhodes about 150 BC and was probably the greatest observational astronomer of antiquity, discovered from his own observations and those of others made over the previous 150 years that the equinoxes, where the ecliptic (the Sun's apparent path) crosses the celestial equator (the celestial equivalent of the terrestrial Equator), were not fixed in space but moved slowly in a westerly direction. The movement is small, amounting to no more than 2° in 150 years, and it is known now as the precession of the equinoxes. Calendrically, it was an important discovery because the tropical year is measured with reference to the equinoxes, and precession reduced the value accepted by Callippus. Hipparchus calculated the tropical year to have a length of 365.242 days, which was very close to the present calculation of 365.242199 days; he also computed the precise length of a lunation, using a "great year" of four Callippic cycles. He arrived at the value of 29.53058 days for a lunation, which, again, is comparable with the present-day figure, 29.53059 days.

The calendar dating of historical events and the determination of how many days have elapsed since some astronomical or other occurrence are difficult for a number of reasons. Leap years have to be inserted, but, not always regularly, months have changed their lengths and new ones have been added from time to time, years have commenced on varying dates and their lengths have been computed in various ways. Since historical dating must take all these factors into account, it occurred to the 16th-century French classicist and literary scholar Joseph Justus Scaliger (1540–1609) that a consecutive numbering system could be of inestimable help. This he thought should be arranged as a cyclic period of great length, and he worked out the system that is known as the Julian period, in honour of his father Julius Caesar Scaliger (1484–1558). He published his proposals in Paris in 1583 under the title *De Emendatione Temporum.*

The Julian period is a cycle of 7,980 years. It is based on the Metonic cycle of 19 years, a "solar cycle" of 28 years, and the Indiction cycle of 15 years. The so-called solar cycle was a period after which the days of the seven-day week repeated on the same dates. Since one year contains 52 weeks of seven days, plus one day, the days of the week would repeat every seven years were no leap year to intervene. A Julian calendar (see below) leap year cycle is four

Cycle of Indiction

years, therefore the days of the week repeat on the same dates every 4 × 7 = 28 years. The cycle of the Indiction was a fiscal, not an astronomical, period. It first appears in tax receipts for Egypt in AD 303, and probably took its origin in a periodic 15-year taxation census that followed Diocletian's reconquest of Egypt in AD 297. By multiplying the Metonic, solar, and Indiction cycles together, Scaliger obtained his cycle of 7,980 years (19 × 28 × 15 = 7,980), a period of sufficient length to cover most previous and future historical dates required at any one time.

Scaliger, tracing each of the three cycles back in time, found that all coincided in the year 4713 BC, on the Julian calendar reckoning. On the information available to him, he believed this to be a date considerably before any historical events. He therefore set the beginning of the first Julian period at January 1, 4713 BC. The years of the Julian period are not now used, but the day number is still used in astronomy and in preparing calendar tables, for it is the only record where days are free from combination into weeks and months.

(C.A.R.)

Ancient and religious calendar systems

THE NEAR EAST AND THE MIDDLE EAST

The lunisolar calendar, in which months are lunar but years are solar—that is, are brought into line with the course of the Sun—was used in the early civilizations of the whole Middle East, except Egypt, and in Greece. The formula was probably invented in Mesopotamia in the 3rd millennium BC. Study of cuneiform tablets found in this region facilitates tracing the development of time reckoning back to the 27th century BC, near the invention of writing. The evidence shows that the calendar is a contrivance for dividing the flow of time into units that suit society's current needs. Though calendar makers put to use time signs offered by nature—the Moon's phases, for example—they rearranged reality to make it fit society's constructions.

Babylonian calendars. In Mesopotamia the solar year was divided into two seasons, the "summer," which included the barley harvest in the second half of May or in the beginning of June, and the "winter," which roughly corresponded to today's fall–winter. Three seasons (Assyria) and four seasons (Anatolia) were counted in northerly countries, but in Mesopotamia the bipartition of the year seemed natural. As late as *c.* 1800 BC the prognoses for the welfare of the city of Mari, on the middle Euphrates, were taken for six months.

The months began at the first visibility of the New Moon, and in the 8th century BC court astronomers still reported this important observation to the Assyrian kings. The names of the months differed from city to city, and within the same Sumerian city of Babylonia a month could have several names, derived from festivals, from tasks (*e.g.,* sheepshearing) usually performed in the given month, and so on, according to local needs. On the other hand, as early as the 27th century BC, the Sumerians had used artificial time units in referring to the tenure of some high official—*e.g.,* on N-day of the turn of office of PN, governor. The Sumerian administration also needed a time unit comprising the whole agricultural cycle; for example, from the delivery of new barley and the settling of pertinent accounts to the next crop. This financial year began about two months after barley cutting. For other purposes, a year began before or with the harvest. This fluctuating and discontinuous year was not precise enough for the meticulous accounting of Sumerian scribes, who by 2400 BC already used the schematic year of 30 × 12 = 360 days.

The royal year

At about the same time, the idea of a royal year took precise shape, beginning probably at the time of barley harvest, when the king celebrated the new (agricultural) year by offering first fruits to gods in expectation of their blessings for the year. When, in the course of this year, some royal exploit (conquest, temple building, and so on) demonstrated that the fates had been fixed favourably by the celestial powers, the year was named accordingly; for example, as the year in which "the temple of Ningirsu was built." Until the naming, a year was described as that "following the year named (after such and such event)." The use of the date formulas was supplanted in Babylonia by the counting of regnal years in the 17th century BC.

The use of lunar reckoning began to prevail in the 21st century BC. The lunar year probably owed its success to economic progress. A barley loan could be measured out to the lender at the next year's threshing floor. The wider use of silver as the standard of value demanded more flexible payment terms. A man hiring a servant in the lunar month of Kislimu for a year knew that the engagement would end at the return of the same month, without counting days or periods of office between two dates. At the city of Mari in about 1800 BC, the allocations were already reckoned on the basis of 29- and 30-day lunar months. In the 18th century BC, the Babylonian Empire standardized the year by adopting the lunar calendar of the Sumerian sacred city of Nippur. The power and the cultural prestige of Babylon assured the success of the lunar year, which began on Nisanu 1, in the spring. When, in the 17th century BC, the dating by regnal years became usual, the period between the accession day and the next Nisanu 1 was described as "the beginning of the kingship of PN," and the regnal years were counted from this Nisanu 1.

It was necessary for the lunar year of about 354 days to be brought into line with the solar (agricultural) year of approximately 365 days. This was accomplished by the use of an intercalated month. Thus, in the 21st century BC, a special name for the intercalated month *iti dirig* appears in the sources. The intercalation was operated haphazardly, according to real or imagined needs, and each Sumerian city inserted months at will; *e.g.,* 11 months in 18 years or two months in the same year. Later, the empires centralized the intercalation, and as late as 541 BC it was proclaimed by royal fiat. Improvements in astronomical knowledge eventually made possible the regularization of intercalation; and, under the Persian kings (*c.* 380 BC), Babylonian calendar calculators succeeded in computing an almost perfect equivalence in a lunisolar cycle of 19 years and 235 months with intercalations in the years 3, 6, 8, 11, 14, 17, and 19 of the cycle. The new year's day (Nisanu 1) now oscillated around the spring equinox within a period of 27 days.

Babylonian months

The Babylonian month names were Nisanu, Ayaru, Simanu, Du'uzu, Abu, Ululu, Tashritu, Arakhsamna, Kislimu, Tebetu, Shabatu, Adaru. The month Adaru II was intercalated six times within the 19-year cycle but never in the year that was 17th of the cycle, when Ululu II was inserted. Thus, the Babylonian calendar until the end preserved a vestige of the original bipartition of the natural year into two seasons, just as the Babylonian months to the end remained truly lunar and began when the New Moon was first visible in the evening. The day began at sunset. Sundials and water clocks served to count hours.

The influence of the Babylonian calendar was seen in many continued customs and usages of its neighbour and vassal states long after the Babylonian Empire had been succeeded by others. In particular, the Jewish calendar in use at relatively late dates employed similar systems of intercalation of months, month names, and other details (see below *The Jewish calendar*). The Jewish adoption of Babylonian calendar customs dates from the period of the Babylonian Exile in the 6th century BC.

Other calendars used in the ancient Near East. *The Assyrians and the Hittites.* Of the calendars of other peoples of the ancient Near East, very little is known. Thus, though the names of all or of some months are known, their order is not. The months were probably everywhere lunar, but evidence for intercalation is often lacking; for instance, in Assyria. For accounting, the Assyrians also used a kind of week, of five days, as it seems, identified by the name of an eponymous official. Thus, a loan could be made and interest calculated for a number of weeks in advance and independently of the vagaries of the civil year. In the city of Ashur, the years bore the name of the official elected for the year; his eponym was known as the *limmu.* As late as about 1070 BC, his installation date was not fixed in the calendar. From about 1100 BC, however, Babylonian month names began to supplant Assyrian names, and, when Assyria became a world power, it used the Babylonian lunisolar calendar.

The calendar of the Hittite Empire is known even less well. As in Babylonia, the first Hittite month was that of first fruits, and, on its beginning, the gods determined the fates.

Iran. At about the time of the conquest of Babylonia in 539 BC, Persian kings made the Babylonian cyclic calendar standard throughout the Persian Empire, from the Indus to the Nile. Aramaic documents from Persian Egypt, for instance, bear Babylonian dates besides the Egyptian. Similarly, the royal years were reckoned in Babylonian style, from Nisanu 1. It is probable, however, that at the court itself the counting of regnal years began with the accession day. The Seleucids and, afterward, the Parthian rulers of Iran maintained the Babylonian calendar. The fiscal administration in northern Iran, from the 1st century BC, at least, used Zoroastrian month and day names in documents in Pahlavi (the Iranian language of Sāsānian Persia). The origin and history of the Zoroastrian calendar year of 12 months of 30 days, plus five days (that is, 365 days), remain unknown. It became official under the Sāsānian dynasty, from about AD 226 until the Arab conquest in 621. The Arabs introduced the Muslim lunar year, but the Persians continued to use the Sāsānian solar year, which in 1079 was made equal to the Julian year by the introduction of the leap year.

he Zoro-
strian
alendar

The Egyptian calendar. The ancient Egyptians originally employed a calendar based upon the Moon, and, like many peoples throughout the world, they regulated their lunar calendar by means of the guidance of a sidereal calendar. They used the seasonal appearance of the star Sirius (Sothis); this corresponded closely to the true solar year, being only 12 minutes shorter. Certain difficulties arose, however, because of the inherent incompatibility of lunar and solar years. To solve this problem the Egyptians invented a schematized civil year of 365 days divided into three seasons, each of which consisted of four months of 30 days each. To complete the year, five intercalary days were added at its end, so that the 12 months were equal to 360 days plus five extra days. This civil calendar was derived from the lunar calendar (using months) and the agricultural, or Nile, fluctuations (using seasons); it was, however, no longer directly connected to either and thus was not controlled by them. The civil calendar served government and administration, while the lunar calendar continued to regulate religious affairs and everyday life.

In time, the discrepancy between the civil calendar and the older lunar structure became obvious. Because the lunar calendar was controlled by the rising of Sirius, its months would correspond to the same season each year, while the civil calendar would move through the seasons because the civil year was about one-fourth day shorter than the solar year. Hence, every four years it would fall behind the solar year by one day, and after 1,460 years it would again agree with the lunisolar calendar. Such a period of time is called a Sothic cycle.

The Sothic
cycle

Because of the discrepancy between these two calendars, the Egyptians established a second lunar calendar based upon the civil year and not, as the older one had been, upon the sighting of Sirius. It was schematic and artificial, and its purpose was to determine religious celebrations and duties. In order to keep it in general agreement with the civil year, a month was intercalated every time the first day of the lunar year came before the first day of the civil year; later, a 25-year cycle of intercalation was introduced. The original lunar calendar, however, was not abandoned but was retained primarily for agriculture because of its agreement with the seasons. Thus, the ancient Egyptians operated with three calendars, each for a different purpose.

The only unit of time that was larger than a year was the reign of a king. The usual custom of dating by reign was: "year 1, 2, 3 . . . , etc., of King So-and-So," and with each new king the counting reverted back to year One. King lists recorded consecutive rulers and the total years of their respective reigns.

The civil year was divided into three seasons, commonly translated: Inundation, when the Nile overflowed the agricultural land; Going Forth, the time of planting when the Nile returned to its bed; and Deficiency, the time of low water and harvest.

The months of the civil calendar were numbered according to their respective seasons and were not listed by any particular name—*e.g.,* third month of Inundation—but for religious purposes the months had names. How early these names were employed in the later lunar calendar is obscure.

The days in the civil calendar were also indicated by number and listed according to their respective months. Thus a full civil date would be: "Regnal year 1, fourth month of Inundation, day 5, under the majesty of King So-and-So." In the lunar calendar, however, each day had a specific name, and from some of these names it can be seen that the four quarters or chief phases of the Moon were recognized, although the Egyptians did not use these quarters to divide the month into smaller segments, such as weeks. Unlike most people who used a lunar calendar, the Egyptians began their day with sunrise instead of sunset because they began their month, and consequently their day, by the disappearance of the old Moon just before dawn.

Designa-
tion of
days

As was customary in early civilizations, the hours were unequal, daylight being divided into 12 parts, and the night likewise; the duration of these parts varied with the seasons. Both water clocks and sundials were constructed with notations to indicate the hours for the different months and seasons of the year. The standard hour of constant length was never employed in ancient Egypt.

(J.D.Sc./C.A.R.)

Ancient Greek calendars in relation to the Middle East.
Earliest sources. The earliest sources (clay tablets of the 13th century BC, the writings of Homer and Hesiod) imply the use of lunar months; Hesiod also uses reckoning determined by the observation of constellations and star groups; *e.g.,* the harvest coincides with the visible rising of the star group known as the Pleiades before dawn. This simultaneous use of civil and natural calendars is characteristic of Greek as well as Egyptian time reckoning. In the classical age and later, the months, named after festivals of the city, began in principle with the New Moon. The lunar year of 12 months and about 354 days was to be matched with the solar year by inserting an extra month every other year. The Macedonians used this system as late as the 3rd century BC, although 25 lunar months amount to about 737 days, while two solar years count about 730 days. In fact, as the evidence from the second half of the 5th century BC shows, at this early time the calendar was already no longer tied in with the phases of the Moon. The cities, rather, intercalated months and added or omitted days at will to adjust the calendar to the course of the Sun and stars and also for the sake of convenience, as, for instance, to postpone or advance a festival without changing its traditional calendar date. The calendric New Moon could disagree by many days with the true New Moon, and in the 2nd century BC Athenian documents listed side by side both the calendar date and that according to the Moon. Thus, the lunar months that were in principle parallel might diverge widely in different cities. Astronomers such as Meton, who in 432 BC calculated a 19-year lunisolar cycle, were not heeded by the politicians, who clung to their calendar-making power.

The year. The civil year (*etos*) was similarly dissociated from the natural year (*eniautos*). It was the tenure term of an official or priest, roughly corresponding to the lunar year, or to six months; it gave his name to his time period. In Athens, for instance, the year began on Hecatombaion 1, roughly midsummer, when the new archon entered his office, and the year was designated by his name; *e.g.,* "when Callimedes was archon"—that is, 360–359 BC. There was no New Year's festival.

Civil and
natural
years
compared

As the archon's year was of indefinite and unpredictable length, the Athenian administration for accounting, for the dates of popular assemblies, etc., used turns of office of the sections (prytanies) of the Council (Boule), which each had fixed length within the year. The common citizen used, along with the civil months, the seasonal time reckoning based on the direct observation of the Moon's phases and on the appearance and setting of fixed stars. A device (called a *parapēgma*) with movable pegs indicated the approximate correspondence between, for example, the rising of the star Arcturus and the civil date.

After Alexander's conquest of the Persian Empire, the Macedonian calendar came to be widely used by the Greeks in the East, though in Egypt it was supplanted by the Egyptian year at the end of the 3rd century BC. The Seleucids, from the beginning, adapted the Macedonian year to the Babylonian 19-year cycle (see above *Babylonian calendars*). Yet, Greek cities clung to their arbitrary system of time reckoning even after the introduction of the Julian calendar throughout the Roman Empire. As late as *c.* AD 200, they used the antiquated *octaëteris* (see above *Complex cycles*).

Months, days, seasons. The Athenian months were called Hecatombaion (in midsummer), Metageitnion, Boedromion, Pyanopsion, Maimacterion, Poseideion, Gamelion, Anthesterion, Elaphebolion, Mounychion, Thargelion, and Scirophorion. The position of the intercalary month varied. Each month, in principle, consisted of 30 days, but in roughly six months the next to last day, the 29th, was omitted. The days were numbered within each of the three decades of the month. Thus, for example, Hecatombaion 16th was called "6th after the 10th of Hecatombaion." The Macedonian months were Dios (in fall), Apellaios, Audynaios, Peritios, Dystros, Xanthicos, Artemisios, Daisios, Panemos, Loos, Gorpiaios, and Hyperberetaios. In the Seleucid calendar, Dios was identified with the Babylonian Tashritu, Apellaios with Arakhsamna, and so on.

Similar to the Babylonian civil pattern, the daylight time and the night were divided into four "watches" and 12 (unequal) hours each. Thus, the length of an hour oscillated between approximately 45 and 75 present-day minutes, according to the season. Water clocks, gnomons, and, after *c.* 300 BC, sundials roughly indicated time. The season division was originally bipartite as in Babylonia—summer and winter—but four seasons were already attested by about 650 BC. (E.J.Bi.)

The early Roman calendar. This originated as a local calendar in the city of Rome, supposedly drawn up by Romulus some seven or eight centuries before the Christian Era. The year began in March and consisted of 10 months, six of 30 days and four of 31 days, making a total of 304 days: it ended in December, to be followed by what seems to have been an uncounted winter gap. Numa Pompilius, according to tradition the second king of Rome (715?–673? BC), is supposed to have added two extra months, January and February, to fill the gap and to have increased the total number of days by 50, making 354. To obtain sufficient days for his new months, he is then said to have deducted one day from the 30-day months, thus having 56 days to divide between January and February. But since the Romans had, or had developed, a superstitious dread of even numbers, January was given an extra day; February was still left with an even number of days, but as that month was given over to the infernal gods, this was considered appropriate. The system allowed the year of 12 months to have 355 days, an uneven number.

The so-called Roman republican calendar was supposedly introduced by the Etruscan Tarquinius Priscus (616–579 BC), according to tradition the fifth king of Rome. He wanted the year to begin in January since it contained the festival of the god of gates (later the god of all beginnings), but expulsion of the Etruscan dynasty in 510 BC led to this particular reform's being dropped. The Roman republican calendar still contained only 355 days, with February having 28 days; March, May, July, and October 31 days each; January, April, June, August, September, November, and December 29 days. It was basically a lunar calendar and short by 10¼ days of a 365¼-day tropical year. In order to prevent it from becoming too far out of step with the seasons, an intercalary month, Intercalans, or Mercedonius (from *merces,* meaning wages, since workers were paid at this time of year), was inserted between February 23 and 24. It consisted of 27 or 28 days, added once every two years, and in historical times at least, the remaining five days of February were omitted. The intercalation was therefore equivalent to an additional 22 or 23 days, so that in a four-year period the total days in the calendar amounted to (4 × 355) + 22 + 23, or 1,465: this gave an average of 366.25 days per year.

Intercalation was the duty of the Pontifices, a board that assisted the chief magistrate in his sacrificial functions. The reasons for their decisions were kept secret, but, because of some negligence and a measure of ignorance and corruption, the intercalations were irregular, and seasonal chaos resulted. In spite of this and the fact that it was over a day too long compared with the tropical year, much of the modified Roman republican calendar was carried over into the Gregorian calendar now in general use. (C.A.R.)

The Jewish calendar. *The calendar in Jewish history.* Present knowledge of the Jewish calendar in use before the period of the Babylonian Exile is both limited and uncertain. The Bible refers to calendar matters only incidentally, and the dating of components of Mosaic Law remains doubtful. The earliest datable source for the Hebrew calendar is the Gezer Calendar, written probably in the age of Solomon, in the late 10th century BC. The inscription indicates the length of main agricultural tasks within the cycle of 12 lunations. The calendar term here is *yereaḥ,* which in Hebrew denotes both "moon" and "month." The second Hebrew term for month, *ḥodesh,* properly means the "newness" of the lunar crescent. Thus, the Hebrew months were lunar. They are not named in pre-exilic sources except in the biblical report of the building of Solomon's Temple in I Kings, where the names of three months, two of them also attested in the Phoenician calendar, are given; the months are usually numbered rather than named. The "beginning of the months" was the month of the Passover (see JUDAISM: *The cycle of the religious year*). In some passages, the Passover month is that of *ḥodesh ha-aviv,* the lunation that coincides with the barley being in the ear. Thus, the Hebrew calendar is tied in with the course of the Sun, which determines ripening of the grain. It is not known how the lunar year of 354 days was adjusted to the solar year of 365 days. The Bible never mentions intercalation. The year *shana,* properly "change" (of seasons), was the agricultural and, thus, liturgical year. There is no reference to the New Year's day in the Bible.

After the conquest of Jerusalem (587 BC), the Babylonians introduced their cyclic calendar (see above *Babylonian calendars*) and the reckoning of their regnal years from Nisanu 1, about the spring equinox. The Jews now had a finite calendar year with a New Year's day, and they adopted the Babylonian month names, which they continue to use. From 587 BC until AD 70, the Jewish civil year was Babylonian, except for the period of Alexander the Great and the Ptolemies (332–200 BC), when the Macedonian calendar was used. The situation after the destruction of the Temple in Jerusalem in AD 70 remains unclear. It is not known whether the Romans introduced their Julian calendar or the calendar that the Jews of Palestine used after AD 70 for their business transactions. There is no calendar reference in the New Testament; the contemporary Aramaic documents from Judaea are rare and prove only that the Jews dated events according to the years of the Roman emperors. The abundant data in the Talmudic sources concern only the religious calendar.

Use of the New Moon

In the religious calendar, the commencement of the month was determined by the observation of the crescent New Moon, and the date of the Passover was tied in with the ripening of barley. The actual witnessing of the New Moon and observing of the stand of crops in Judaea were required for the functioning of the religious calendar. The Jews of the Diaspora, or Dispersion, who generally used the civil calendar of their respective countries, were informed by messengers from Palestine about the coming festivals. This practice is already attested for 143 BC. After the destruction of the Temple in AD 70, rabbinic leaders took over from the priests the fixing of the religious calendar. Visual observation of the New Moon was supplemented and toward AD 200, in fact, supplanted by secret astronomical calculation. But the people of the Diaspora were often reluctant to wait for the arbitrary decision of the calendar makers in the Holy Land. Thus, in Syrian Antioch in AD 328–342, the Passover was always celebrated in (Julian) March, the month of the spring equinox, without regard to the Palestinian rules and rulings. To preserve the unity of Israel, the patriarch Hillel II, in 358/359, published the "secret" of calendar making, which essentially consisted of the use of the Babylonian 19-year cycle with some modifications required by the Jewish ritual.

The application of these principles occasioned controversies as late as the 10th century AD. In the 8th century, the Karaites, following Muslim practice, returned to the actual observation of the crescent New Moon and of the stand of barley in Judaea. But some centuries later they also had to use a precalculated calendar. The Samaritans, likewise, used a computed calendar.

Because of the importance of the Sabbath as a time divider, the seven-day week served as a time unit in Jewish worship and life. As long as the length of a year and of every month remained unpredictable, it was convenient to count weeks. The origin of the biblical septenary, or seven-day, week remains unknown; its days were counted from the Sabbath (Saturday for the Jews and Sunday for Christians). A visionary, probably writing in the Persian or early Hellenistic age under the name of the prediluvian Enoch, suggested the religious calendar of 364 days, or 52 weeks, based on the week, in which all festivals always fall on the same weekday. His idea was later taken up by the Qumrān community.

(E.J.Bi.)

The structure of the calendar. The Jewish calendar in use today is lunisolar, the years being solar and the months lunar, but it also allows for a week of seven days. Because the year exceeds 12 lunar months by about 11 days, a 13th month of 30 days is intercalated in the third, sixth, eighth, 11th, 14th, 17th, and 19th years of a 19-year cycle. For practical purposes—*e.g.,* for reckoning the commencement of the Sabbath—the day begins at sunset; but the calendar day of 24 hours always begins at 6 PM. The hour is divided into 1,080 parts (*ḥalaqim;* this division is originally Babylonian), each part (*ḥeleq*) equalling $3\frac{1}{3}$ seconds. The *ḥeleq* is further divided into 76 *regaʿim.*

'Full" and 'defective" months

The synodic month is the average interval between two mean conjunctions of the Sun and Moon, when these bodies are as near as possible in the sky, which is reckoned at 29 days 12 hours 44 minutes $3\frac{1}{3}$ seconds; a conjunction is called a *molad.* This is also a Babylonian value. In the calendar month, however, only complete days are reckoned, the "full" month containing 30 days and the "defective" month 29 days. The months Nisan, Sivan (Siwan), Av, Tishri, Shevaṭ, and, in a leap year, First Adar are always full; Iyyar, Tammuz, Elul, Ṭevet, and Adar (known as Second Adar, or Adar Sheni, in a leap year) are always defective, while Ḥeshvan (Ḥeshwan) and Kislev (Kislew) vary. The calendar, thus, is schematic and independent of the true New Moon. The number of days in a year varies. The number of days in a synodic month multiplied by 12 in a common year and by 13 in a leap year would yield fractional figures. Hence, again reckoning complete days only, the common year has 353, 354, or 355 days and the leap year 383, 384, or 385 days. A year in which both Ḥeshvan and Kislev are full, called complete (*shelema*), has 355 or (if a leap year) 385 days; a normal (*sedura*) year, in which Ḥeshvan is defective and Kislev full, has 354 or 384 days; while a defective (*ḥasera*) year, in which both these months are defective, has 353 or 383 days. The character of a year (*qeviʿa,* literally "fixing") is described by three letters of the Hebrew alphabet, the first and third giving, respectively, the days of the weeks on which the New Year occurs and Passover begins, while the second is the initial of the Hebrew word for defective, normal, or complete. There are 14 types of *qeviʿot,* seven in common and seven in leap years. The New Year begins on Tishri 1, which may be the day of the *molad* of Tishri but is often delayed by one or two days for various reasons. Thus, in order to prevent the Day of Atonement (Tishri 10) from falling on a Friday or a Sunday and the seventh day of Tabernacles (Tishri 21) from falling on a Saturday, the New Year must avoid commencing on Sundays, Wednesdays, or Fridays. Again, if the *molad* of Tishri occurs at noon or later, the New Year is delayed by one or, if this would cause it to fall as above, two days. These delays (*deḥiyyot*) necessitate, by reason of the above-mentioned limits on the number of days in the year, two other delays.

The seasons in the Jewish calendar

The mean beginning of the four seasons is called *tequfa* (literally "orbit," or "course"); the *tequfa* of Nisan denotes the mean Sun at the vernal equinox, that of Tammuz at the summer solstice, that of Tishri at the autumnal equinox, and that of Ṭevet at the winter solstice. As 52 weeks are the equivalent to 364 days, and the length of the solar year is nearly $365\frac{1}{4}$ days, the *tequfot* move forward in the week by about $1\frac{1}{4}$ days each year. Accordingly, reckoning the length of the year at the approximate value of $365\frac{1}{4}$ days, they are held to revert after 28 years ($28 \times 1\frac{1}{4} = 35$ days) to the same hour on the same day of the week (Tuesday, 6 PM) as at the beginning. This cycle is called the great, or solar, cycle (*maḥzor gadol* or *ḥamma*). The present Jewish calendar is mainly based on the more accurate value, in excess of the true tropical year by about 6 minutes 40 seconds. Thus, it is advanced by one day in about 228 years with regard to the equinox.

To a far greater extent than the solar cycle of 28 years, the Jewish calendar employs, as mentioned above, a small, or lunar, cycle (*maḥzor qaṭan*) of 19 years, adjusting the lunar months to the solar years by intercalations. Passover, on Nisan 14, is not to begin before the spring *tequfa,* and so the intercalary month is added after Adar. The *maḥzor qaṭan* is akin to the Metonic cycle described above.

The Jewish Era in use today is that dated from the supposed year of the Creation (designated *anno mundi* or AM) with its epoch, or beginning, in 3761 BC. For example, the Jewish year 5745 AM, the 7th in the 303rd lunar cycle and the 5th in the 206th solar cycle, is a regular year of 12 months, or 354 days. The *qeviʿa* is, using the three respective letters of the Hebrew alphabet as two numerals and an initial in the manner indicated above, HKZ, which indicates that Rosh Hashana (New Year) begins on the fifth (H = 5) and Passover on the seventh (Z = 7) day of the week and that the year is regular (K = *ke-sidra*); *i.e.,* Ḥeshvan is defective—29 days, and Kislev full—30 days. The Jewish year 5745 AM corresponds with the Christian Era period that began September 27, 1984, and ended September 15, 1985. Neglecting the thousands, current Jewish years AM are converted into years of the current Christian Era by adding 239 or 240—239 from the Jewish New Year (about September) to December 31 and 240 from January 1 to the eve of the Jewish New Year. The adjustment differs slightly for the conversion of dates of now-antiquated versions of the Jewish Era of the Creation and the Christian Era, or both. Tables for the exact conversion of such dates are available.

Months and important days. The months of the Jewish year and the notable days are as follows:

Tishri: 1–2, Rosh Hashana (New Year); 3, Fast of Gedaliah; 10, Yom Kippur (Day of Atonement); 15–21, Sukkot (Tabernacles); 22, Shemini Atzeret (Eighth Day of Solemn Assembly); 23, Simḥat Torah (Rejoicing of the Law).
Ḥeshvan.
Kislev: 25, Ḥanukka (Festival of Lights) begins.
Tevet: 2 or 3, Ḥanukka ends; 10, Fast.
Shevaṭ: 15, New Year for Trees (Mishna).
Adar: 13, Fast of Esther; 14–15, Purim (Lots).
Second Adar (Adar Sheni) or ve-Adar (intercalated month); Adar holidays fall in ve-Adar during leap years.
Nisan: 15–22, Pesaḥ (Passover).
Iyyar: 5, Israel Independence Day.
Sivan: 6–7, Shavuot (Feast of Weeks [Pentecost]).
Tammuz: 17, Fast (Mishna).
Av: 9, Fast (Mishna).
Elul.

(E.J.Wi.)

The Muslim calendar. The Muslim Era is computed from the starting point of the year of the emigration (Hegira); that is, from the year in which Muḥammad, the prophet of Islām, emigrated from Mecca to Medina, AD 622. The second caliph, ʿUmar I, who reigned 634–644, set the first day of the month Muḥarram as the beginning of the year; that is, July 16, 622, which had already been fixed by the Qurʾan as the first day of the year.

The years of the Muslim calendar are lunar and always consist of 12 lunar months alternately 30 and 29 days long, beginning with the approximate New Moon. The year has 354 days, but the last month (Dhū al-Ḥijjah) sometimes has an intercalated day, bringing it up to 30 days and making a total of 355 days for that year. The months do not keep to the same seasons in relation to the Sun, because there are no intercalations of months. The months regress through all the seasons every 32½ years.

Ramaḍān

Ramaḍān, the ninth month, is observed throughout the Muslim world as a month of fasting. According to the Qur'an, Muslims must see the New Moon with the naked eye before they can begin their fast. The practice has arisen that two witnesses should testify to this before a *qaḍī* (judge), who, if satisfied, communicates the news to the *muftī* (the interpreter of Muslim law), who orders the beginning of the fast. It has become usual for Middle Eastern Arab countries to accept, with reservations, the verdict of Cairo. Should the New Moon prove to be invisible, then the month Shaʿbān, immediately preceding Ramaḍān, will be reckoned as 30 days in length, and the fast will begin on the day following the last day of this month. The end of the fast follows the same procedure.

The era of the Hegira is the official era in Saudi Arabia, Yemen, and the principalities of the Persian Gulf. Egypt, Syria, Jordan, and Morocco use both the Muslim and the Christian eras. In all Muslim countries, people use the Muslim Era in private, even though the Christian Era may be in official use.

Some Muslim countries have made a compromise on this matter. Turkey, as early as 1088 AH (AD 1677), took over the solar (Julian) year with its month names but kept the Muslim Era. March 1 was taken as the beginning of the year (commonly called *marti year,* after the Turkish word *mart,* for March). Late in the 19th century, the Gregorian calendar was adopted. In the 20th century, Pres. Mustafa Kemal Atatürk ordered a complete change to the Christian Era. Iran, under Reza Shah Pahlavi (reigned 1925–41), also adopted the solar year but with Persian names for the months and keeping the Muslim Era. March 21 is the beginning of the Iranian year. Thus, the Iranian year 1359 began on March 21, 1980. This era is still in use officially. (See also ISLĀM: *Sacred places and days.*)

(N.A.Z.)

THE FAR EAST

The Hindu calendar. While the Republic of India has adopted the Gregorian calendar for its secular life, its Hindu religious life continues to be governed by the traditional Hindu calendar. This calendar, based primarily on the lunar revolutions, is adapted to solar reckoning.

Early history. The oldest system, in many respects the basis of the classical one, is known from texts of about 1000 BC. It divides an approximate solar year of 360 days into 12 lunar months of 27 (according to the early Vedic text Taittirīya Saṃhitā 4.4.10.1–3) or 28 (according to the Atharvaveda, the fourth of the Vedas, 19.7.1.) days. The resulting discrepancy was resolved by the intercalation of a leap month every 60 months. Time was reckoned by the position marked off in constellations on the ecliptic in which the Moon rises daily in the course of one lunation (the period from New Moon to New Moon) and the Sun rises monthly in the course of one year. These constellations (*nakṣatra*) each measure an arc of 13° 20′ of the ecliptic circle. The positions of the Moon were directly observable, and those of the Sun inferred from the Moon's position at Full Moon, when the Sun is on the opposite side of the Moon. The position of the Sun at midnight was calculated from the *nakṣatra* that culminated on the meridian at that time, the Sun then being in opposition to that *nakṣatra.* The year was divided into three thirds of four months, each of which would be introduced by a special religious rite, the *cāturmāsya* (four-month rite). Each of these periods was further divided into two parts (seasons or *ṛtu*): spring (*vasanta*), from mid-March until mid-May; summer (*grīṣma*), from mid-May until mid-July; the rains (*varṣa*), from mid-July until mid-September; autumn (*śarad*), from mid-September until mid-November; winter (*hemanta*), from mid-November until mid-January; and the dews (*śiśira*), from mid-January until mid-March. The spring months in early times were Madhu and Mādhava, the summer months Śukra and Śuci, the rainy months Nabhas and Nabhasya, the autumn months Īṣa and Ūrja, the winter months Sahas and Sahasya, and the dewy months Tapas and Tapasya. The month, counted from Full Moon to Full Moon, was divided into two halves (*pakṣa,* "wing") of waning (*kṛṣṇa*) and waxing (*śukla*) Moon, and a special ritual (*darśapūrṇamāsa,* "new and full moon rites") was prescribed on the days of New Moon (*amāvasya*) and Full Moon (*pūrṇimās*). The month had theoretically 30 days (*tithi;* see below), and the day (*divasa*) 30 hours (*muhūrta*).

The *nakṣatra*

This picture is essentially confirmed by the first treatise on time reckoning, the *Jyotiṣa-vedāṅga* ("Vedic auxiliary [text] concerning the luminaries") of about 100 BC, which adds a larger unit of five years (*yuga*) to the divisions. A further old distinction is that of two year moieties, the *uttarāyaṇa* ("northern course"), when the Sun has passed the spring equinox and rises every morning farther north, and the *dakṣiṇāyana* ("southern course"), when it has passed the autumnal equinox and rises progressively farther south.

The classical calendar. In its classic form (*Sūrya-siddhānta,* 4th century AD) the calendar continues from the one above with some refinements. With the influence of Hellenism, Greek and Mesopotamian astronomy and astrology were introduced. Though astronomy and time reckoning previously were dictated by the requirements of rituals, the time of which had to be fixed correctly, and not for purposes of divination, the new astrology came into vogue for casting horoscopes and making predictions. Zodiacal time measurement was now used side by side with the older *nakṣatra* one. The *nakṣatra* section of the ecliptic (13° 20′) was divided into four parts of 3° 20′ each; thus, two full *nakṣatra*s and a quarter of one make up one zodiac period, or sign (30°). The year began with the entry of the Sun (*saṃkrānti*) in the sign of Aries. The names of the signs (*rāśi*) were taken over and mostly translated into Sanskrit: *meṣa* ("ram," Aries), *vṛṣabha* ("bull," Taurus), *mithuna* ("pair," Gemini), *karkaṭa* ("crab," Cancer), *siṃha* ("lion," Leo), *kanyā* ("maiden," Virgo), *tulā* ("scale," Libra), *vṛścika* ("scorpion," Scorpius), *dhanus* ("bow," Sagittarius), *makara* ("crocodile," Capricornus), *kumbha* ("water jar," Aquarius), *mīna* ("fish," Pisces).

The precession of the vernal equinox from the Sun's entry into Aries to some point in Pisces, with similar consequences for the summer solstice, autumnal equinox, and winter solstice, has led to two different methods of calculating the *saṃkrānti* (entry) of the Sun into a sign. The precession (*ayana*) is not accounted for in the *nirayana* system (without *ayana*), which thus dates the actual *saṃkrānti* correctly but identifies it wrongly with the equinox or solstice, and the *sāyana* system (with *ayana*), which thus dates the equinox and solstice correctly but identifies it wrongly with the *saṃkrānti.*

While the solar system has extreme importance for astrology, which, it is claimed, governs a person's life as an individual or part of a social system, the sacred time continues to be reckoned by the lunar *nakṣatra* system. The lunar day (*tithi*), a 30th part of the lunar month, remains the basic unit. Thus, as the lunar month is only about 29½ solar days, the *tithi* does not coincide with the natural day (*ahorātra*). The convention is that that *tithi* is in force for the natural day that happened to occur at the dawn of that day. Therefore, a *tithi* beginning after dawn one day and expiring before dawn the next day is eliminated, not being counted in that month, and there is a break in the day sequence.

The names of the *nakṣatra*s, to which correspond the *tithi*s in the monthly lunar cycle and segments of months in the annual solar cycle, are derived from the constellations on the horizon at that time and have remained the same. The names of the months have changed: Caitra (March-April), Vaiśākha (April-May), Jyaiṣṭha (May-June), Āṣāḍha (June-July), Śrāvaṇa (July-August), Bhādrapada (August-September), Āśvina (September-October), Kārttika (October-November), Mārgaśīrṣa (November-December), Pauṣa (December-January), Māgha (January-February), and Phālguna (February-March).

Month names

In this calendar the date of an event takes the following form: month, fortnight (either waning or waxing Moon), name (usually the number) of the *tithi* in that fortnight, and the year of that era which the writer follows. Identification, particularly of the *tithi,* is often quite complicated, since it requires knowledge of the time of sunrise on that day and which 30th of the lunar month was in force then.

Eventually, India also adopted the seven-day week (*saptāha*) from the West and named the days after the corresponding planets: Sunday after the Sun, *ravivāra;* Monday after the Moon, *somavāra;* Tuesday after Mars, *maṅgalavāra;* Wednesday after Mercury, *budhavāra;* Thursday after Jupiter, *bṛhaspativāra;* Friday after Venus, *śukravāra;* and Saturday after Saturn, *śanivāra.*

A further refinement of the calendar was the introduction into dating of the place of a year according to its position in relation to the orbital revolution of the planet Jupiter, called *bṛhaspati* in Sanskrit. Jupiter has a sidereal period (its movement with respect to the "fixed" stars) of 11 years, 314 days, and 839 minutes, so in nearly 12 years it is back into conjunction with those stars from which it began its orbit. Its synodic period brings it into conjunction with the Sun every 398 days and 88 minutes, a little more than a year. Thus, Jupiter in a period of almost 12 years passes about the same series of *nakṣatra*s that the Sun passes in one year and, in a year, about the same *nakṣatra*s as the Sun in a month. A year then can be dated as the month of a 12-year cycle of Jupiter, and the date is given as, for example, grand month of Caitra. This is extended to a unit of five cycles, or the 60-year cycle of Jupiter (*bṛhaspaticakra*), and a "century" of 60 years is formed. This system is known from the 6th century AD onward.

At the other end of the scale, more precision is brought to the day. Every *tithi* is divided into two halves, called *karaṇa*s. The natural day is divided into units ranging from a *vipala* (0.4 second) to a *ghaṭikā* (24 minutes) and an "hour" (*muhūrta*) of 48 minutes; the full natural day has 30 such hours. The day starts at dawn; the first six *ghaṭikā*s are early morning, the second set of six midmorning, the third midday, the fourth afternoon, the fifth evening. Night lasts through three units (*yāma*) of time: six *ghaṭikā*s after sundown, or early night; two of midnight; and four of dawn.

Solar holidays

The sacred calendar. There are a few secular state holidays (*e.g.,* Independence Day) and some solar holidays, such as the entry of the Sun into the sign of Aries (*meṣa-saṃkrānti*), marking the beginning of the new astrological year; the Sun's entry into the sign of Capricornus (*makara-saṃkrānti*), which marks the winter solstice but has coalesced with a hoary harvest festival, which in southern India is very widely celebrated as the Poṅgal festival; and the *mahāviṣuva* day, which is New Year's Eve. But all other important festivals are based on the lunar calendar. As a result of the high specialization of deities and events celebrated in different regions, there are hundreds of such festivals, most of which are observed in smaller areas, though some have followings throughout India. A highly selective list of the major ones, national and regional, follows.

Rāmanavamī ("ninth of Rāma"), on Caitra Ś. (= *śukla,* "waxing fortnight") 9, celebrates the birth of Rāma.

Rathayātrā ("pilgrimage of the chariot"), Āṣāḍha Ś. 2, is the famous Juggernaut (Jagannātha) festival of the temple complex at Puri, Orissa.

Janmāṣṭamī ("eighth day of the birth"), Śrāvaṇa K. (= *kṛṣṇa,* "waning fortnight") 8, is the birthday of the god Kṛṣṇa.

Gaṇeśacaturthī ("fourth of Gaṇeśa"), Bhādrapada Ś. 4, is observed in honour of the elephant-headed god Gaṇeśa, a particular favourite of Mahārāshtra.

Durgā-pūjā ("homage to Durgā"), Āśvina Ś. 7–10, is special to Bengal, in honour of the destructive and creative goddess Durgā.

Daśahrā ("ten days"), or Dussera, Āśvina 7–10, is parallel to Durgā-pūjā, celebrating Rāma's victory over Rāvaṇa, and traditionally the beginning of the warring season.

Lakṣmīpūjā ("homage to Lakṣmī"), Āśvina Ś. 15, is the date on which commercial books are closed, new annual records begun, and business paraphernalia honoured; for Lakṣmī is the goddess of good fortune.

Dīpāvalī, Dīwālī ("strings of lights"), Kārttika K. 15 and Ś. 1, is the festival of lights, when light is carried from the waning to the waxing fortnight and presents are exchanged.

Mahā-śivarātrī ("great night of Śiva"), Māgha K. 13, is when the dangerous but, if placated, benevolent god Śiva is honoured on the blackest night of the month.

Holī (name of a demoness), Phālguna S. 14, is a fertility and role-changing festival, scene of great fun-poking at superiors.

Dolāyātrā ("swing festival"), Phālguna S. 15, is the scene of the famous hook-swinging rites of Orissa.

Gurū Nānak Jayantī, Kārttika S. 15, is the birthday of Nānak, the founder of the sect of Sikhism.

The eras. Not before the first century BC is there any evidence that the years of events were recorded in well-defined eras, whether by cycles, as the Olympic Games in Greece and the tenures of consuls in Rome, or the Roman year dating from the foundation of the city. Perhaps under outside influence, the recording of eras was begun at various times, but these were without universal appeal, and few have remained influential. Among those are (1) the Vikrama era, begun 58 BC; (2) the Śaka era, begun AD 78 (these two are the most commonly used); (3) the Gupta era, begun AD 320; (4) the Harṣa era, begun AD 606. All these were dated from some significant historical event. Of more mythological interest is the Kali era (Kali being the latest and most decadent period in the system of the four *Yuga*s), which is thought to have started either at dawn on February 18, 3102 BC, or at midnight between February 17 and 18 in that year. (J.A.B.v.B.)

The Chinese calendar. Evidence from the Shang oracle bone inscriptions shows that at least by the 14th century BC the Shang Chinese had established the solar year at 365¼ days and lunation at 29½ days. In the calendar that the Shang used, the seasons of the year and the phases of the Moon were all supposedly accounted for. One of the two methods that they used to make this calendar was to add an extra month of 29 or 30 days, which they termed the 13th month, to the end of a regular 12-month year. There is also evidence that suggests that the Chinese developed the Metonic cycle (see above *Complex cycles*)—*i.e.,* 19 years with a total of 235 months—a century ahead of Meton's first calculation (no later than the Spring and Autumn period, 770–476 BC). During this cycle of 19 years there were seven intercalations of months. The other method, which was abandoned soon after the Shang started to adopt it, was to insert an extra month between any two months of a regular year. Possibly, a lack of astronomical and arithmetical knowledge allowed them to do this.

Names of the points

By the 3rd century BC, the first method of intercalation was gradually falling into disfavour, while the establishment of the meteorological cycle, the *erh-shih-ssu chieh-ch'i* (Pinyin *ershisi jieqi*), during this period officially revised the second method. This meteorological cycle contained 24 points, each beginning one of the periods named consecutively the Spring Begins, the Rain Water, the Excited Insects, the Vernal Equinox, the Clear and Bright, the Grain Rains, the Summer Begins, the Grain Fills, the Grain in Ear, the Summer Solstice, the Slight Heat, the Great Heat, the Autumn Begins, the Limit of Heat, the White Dew, the Autumn Equinox, the Cold Dew, the Hoar Frost Descends, the Winter Begins, the Little Snow, the Heavy Snow, the Winter Solstice, the Little Cold, and the Severe Cold. The establishment of this cycle required a fair amount of astronomical understanding of the Earth as a celestial body, and without elaborate equipment it is impossible to collect the necessary information. Modern scholars acknowledge the superiority of pre-Sung Chinese astronomy (at least until about the 13th century AD) over that of other, contemporary nations.

The 24 points within the meteorological cycle coincide with points 15° apart on the ecliptic (the plane of the Earth's yearly journey around the Sun or, if it is thought that the Sun turns around the Earth, the apparent journey of the Sun against the stars). It takes about 15.2 days for the Sun to travel from one of these points to another (because the ecliptic is a complete circle of 360°), and the Sun needs $365\frac{1}{4}$ days to finish its journey in this cycle. Supposedly, each of the 12 months of the year contains two points, but, because a lunar month has only $29\frac{1}{2}$ days and the two points share about 30.4 days, there is always the chance that a lunar month will fail to contain both points, though the distance between any two given points is only 15°. If such an occasion occurs, the intercalation of an extra month takes place. For instance, one may find a year with two "Julys" or with two "Augusts" in the Chinese calendar. In fact, the exact length of the month in the Chinese calendar is either 30 days or 29 days—a phenomenon which reflects its lunar origin. Also, the meteorological cycle means essentially a solar year. The Chinese thus consider their calendar as *yin-yang li,* or a "lunar–solar calendar."

The *yin-yang li*

Although the *yin-yang li* has been continuously employed by the Chinese, foreign calendars were introduced to the Chinese, the Hindu calendar, for instance, during the T'ang (Tang) dynasty (618–907), and were once used concurrently with the native calendar. This situation also held true for the Muslim calendar, which was introduced during the Yüan dynasty (1206–1368). The Gregorian calendar was taken to China by Jesuit missionaries in 1582, the very year that it was first used by Europeans. Not until 1912, after the general public adopted the Gregorian calendar, did the *yin-yang li* lose its primary importance.

One of the most distinguished characteristics of the Chinese calendar is its time-honoured day-count system. By combining the 10 celestial stems, *kan* (*gan*), and the 12 terrestrial branches, *chih* (*zhi*), into 60 units, the Shang Chinese counted the days with *kan-chih* (*gan-zhi*) combinations cyclically. For more than 3,000 years, no one has ever tried to discard the *kan-chih* day-count system. Out of this method there developed the idea of *hsün,* (*xun*), 10 days, which some scholars would render into English as "week." The *kan-chih* combinations probably were adopted for year count by Han emperors during the 2nd century AD.

The *yin-yang li* may have been preceded by a pure lunar calendar because there is one occurrence of the "14th month" and one occurrence of the "15th month" in the Shang oracle bone inscriptions. Unless there was a drastic change in the computation, it is quite inconceivable that an extra 90 days should have been added to a regular year. Julius Caesar had made 45 BC into a year of 445 days for the sake of the adoption of the Julian calendar in the next year. Presumably, the Shang king could have done the same for similar reasons. From the above discussion on the intercalation of months, it is clear that within the *yin-yang li* the details of the lunar calendar are more important than those of the solar calendar. In a solar calendar the 24 meteorological points would recur on the same days every year. Moreover, if a solar calendar were adopted first, then the problem of intercalation would be more related to the intercalation of days rather than intercalation of months.

Many traditional Chinese scholars tried to synchronize the discrepancy between the lunation and the solar year. Some even developed their own ways of computation embodying accounts of eclipses and of other astronomical phenomena. These writings constitute the bulk of the traditional almanacs. In the estimation of modern scholars, at least 102 kinds of almanacs were known, and some were used regularly. The validity or the popularity of each of these almanacs depends heavily on the author's proficiency in handling planetary cycles. In the past these authors competed with one another for the position of calendar master in the Imperial court, even though mistakes in their almanacs could bring them punishment, including death.

(Ch.L.)

THE AMERICAS

The Mayan calendar. The basic structure of the Mayan calendar is common to all calendars of Meso-America (*i.e.,* the civilized part of ancient Middle America). It consists of a ritual cycle of 260 named days and a year of 365 days. These cycles, running concurrently, form a longer cycle of 18,980 days, or 52 years of 365 days, called a "Calendar Round," at the end of which a designated day recurs in the same position in the year.

The "Calendar Round"

The native Mayan name for the 260-day cycle is unknown. Some authorities call it the Tzolkin (Count of Days); others refer to it as the Divinatory Calendar, the Ritual Calendar, or simply the day cycle. It is formed by the combination of numerals 1 through 13, meshing day by day with an ordered series of 20 names. The names of the days differ in the languages of Meso-America, but there is enough correspondence of meaning to permit the correlation of the known series, and there is reason to think that all day cycles were synchronous. The days were believed to have a fateful character, and the Tzolkin was used principally in divination. Certain passages in the *Dresden Codex,* one of the three Mayan manuscripts that survived the conquest, show various Tzolkins divided into four parts of 65 days each, or into five parts of 52 days. The parts are in turn subdivided into a series of irregular intervals, and each interval is accompanied by a group of hieroglyphs and by an illustration, usually depicting a deity performing some simple act. The hieroglyphs apparently give a prognostication, but just how the Maya determined the omens is not known.

The 365-day year was divided into 18 named months of 20 days each, with an additional five days of evil omen, named Uayeb. In late times, the Maya named the years after their first days. Since both the year and the number (20) of names of days are divisible by five, only four names combined with 13 numbers could begin the year. These were called Year Bearers and were assigned in order to the four quarters of the world with their four associated colours. Unlike day cycles, years were not synchronous in all regions. They began at different times and in different seasons, and even among Maya-speaking peoples there was imperfect concordance of the months. Some differences may be due to postconquest attempts to keep the native year in step with the Christian calendar; others no doubt have an earlier origin.

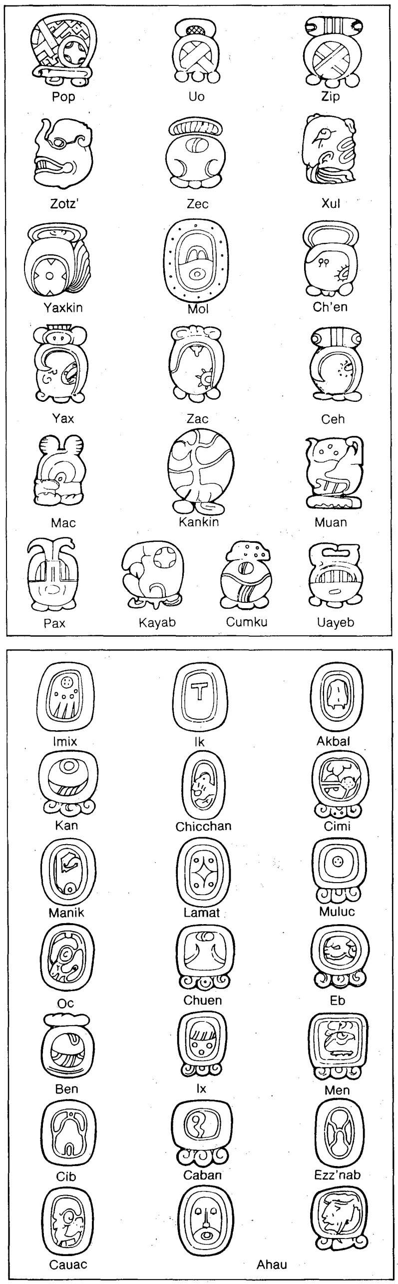
The manner of recording historical dates is peculiar to the ancient Mayan calendar. The Maya did not use the names of years for this purpose. To identify a date of the Calendar Round, they designated the day by its numeral and name, and added the name of the current month, indicating the number of its days that had elapsed by prefixing one of the numerals from 0 through 19. A date written in this way will occur once in every Calendar Round, at intervals of 52 years.

This was not good enough to link events over longer periods of time. Mayan interest in history, genealogy, and astrology required accurate records of events far in the past. To connect dates to one another, the Maya expressed distances between them by a count of days and their multiples. They used what was essentially a vigesimal place-value system of numeration, which is one based on a count of 20, but modified it by substituting 18 for 20 as the multiplier of units of the second order, so that each unit in the third place had the value of 360 days instead of 400. In monumental inscriptions, the digits are usually accompanied by the names of the periods their units represent, although in the manuscripts the period names are omitted and placement alone indicates the value of the units. The period names in ascending order are: *kin* (day); *uinal* (20 days); *tun* (18 *uinals* or 360 days); *katun* (20 *tuns* or 7,200 days); *baktun* or cycle—native name unknown—(20 *katuns* or 144,000 days); and so on up to higher periods. By introducing an odd multiplier to form the *tun,* the Maya made multiplication and division difficult, and there are in the *Dresden Codex* long tables of multiples of numbers that could be more simply manipulated by addition and subtraction.

Period names

To correlate all historical records and to anchor dates firmly in time, the Maya established the "Long Count," a continuous count of time from a base date, 4 Ahau 8 Cumku, which completed a round of 13 *baktuns* far in the past. There were several ways in which one could indicate the position of a Calendar Round dated in the Long Count. The most direct and unambiguous was to use an Initial-Series (IS) notation. The series begins with an outsized composition of signs called the Initial-Series-introducing glyph, which is followed by a count of periods written in descending order. On the earliest known monument, Stela 29 from Tikal in Guatemala, the Initial Series reads: 8 *baktuns,* 12 *katuns,* 14 *tuns,* 8 *uinals,* 15 *kins,* which is written: IS. 8.12.14.8.15. It shows that the Calendar Round date that follows falls 1,243,615 days (just under 3,405 years) after the 4 Ahau 8 Cumku on which the Long Count is based. Stela 29 is broken, and its Calendar Round date is missing, but from the information above, it can be calculated to have been 13 Men 3 Zip (the 195th day of the Tzolkin, the 44th of the year).

By courtesy of the trustees of the British Museum



The Mayan calendar.
(Top) Month signs. (Bottom) Day-signs.

Normally, only the opening date of an inscription is written as an Initial Series. From this date, distance numbers, called Secondary Series (SS), lead back or forward to other dates in the record, which frequently ends with a Period-Ending (PE) date. This is a statement that a given date completes a whole number of *tuns* or *katuns* in the next higher period of the Long Count. Such a notation identifies the date unambiguously within the historic period. The latest Period Ending recorded on a given monument is also known as its Dedicatory Date (DD), for it was a common custom to set up monuments on the completion of *katuns* of the Long Count and sometimes also at the end of every five or 10 *tuns.* The Maya also celebrated *katun* and five-*tun* "anniversaries" of important dates and recorded them in much the same way as the period endings.

Period-Ending dates gradually took the place of Initial Series, and, in northern Yucatán, where Mayan sites of the latest period are located, a new method of notation dispensed with distance numbers altogether by noting after a Calendar Round date the number of the current *tun* in a Long Count *katun* named by its last day. Long Count *katuns* end with the name Ahau (Lord), combined with one of 13 numerals; and their names form a Katun Round of 13 *katuns.* This round is portrayed in Spanish colonial manuscripts as a ring of faces depicting the Lords. There are also recorded prophesies for *tuns* and *katuns,* which make many allusions to history, for the Maya seem to have conceived time, and even history itself, as a series of cyclical, recurring events.

The discontinuance of Initial-Series notations some centuries before the conquest of Mexico by Spain makes all attempted correlations of the Mayan count with the Christian calendar somewhat uncertain, for such correlations are all based on the assumption that the Katun Round of early colonial times was continuous with the ancient Long Count. The correlation most in favour now equates the 4 Ahau 8 Cumku that begins the Mayan count with the Julian day 584,283 (see above *Complex cycles*). According to this correlation, the *katun* 13 Ahau that is said to have ended shortly before the foundation of Mérida, Yucatán, ended on November 14, 1539, by the Gregorian calendar, and it was the Long Count *katun* 11.16.0.0.0. 13 Ahau 8 Xul. Some tests of archaeological material by the radiocarbon method corroborate this correlation; but results are not sufficiently uniform to resolve all doubts, and some archaeologists would prefer to place the foundation of Mérida in the neighbourhood of 12.9.0.0.0. in the Mayan count. Correlations based on astronomical data so far have been in conflict with historical evidence, and none has gained a significant degree of acceptance.

Lunar correlations

The basic elements of the Mayan calendar have little to do with astronomy. A lunar count was, however, included in a Supplementary Series appended to Initial-Series dates. The series is composed of hieroglyphs labelled Glyphs *G, F, E* or *D, C, B,* and *A,* and a varying number of others. Glyph *G* changes its form daily, making a round of nine days, possibly corresponding to the nine gods of the night hours or Mexican Lords of the Night. Glyph *F* is closely associated with Glyph *G* and does not vary. Glyphs *E* and *D* have numerical coefficients that give the age of the current Moon within an error of two or three days; Glyph *C* places it in a lunar half year; and Glyph *A* shows whether it is made up of 29 or 30 days. The meaning of Glyph *B* is unknown. There are discrepancies in the lunar records from different sites, but during a period of about 80 years, called the Period of Uniformity, a standard system of grouping six alternating 29- and 30-day moons was used everywhere.

Occasionally included with the Supplementary Series is a date marking the conclusion of an 819-day cycle shortly before the date of Initial Series. The number of days in this cycle is obtained by multiplying together 13, 9, and 7, all very significant numbers in Mayan mythology.

It has been suggested that certain other dates, called determinants, indicate with a remarkable degree of accuracy how far the 365-day year had diverged from the solar year since the beginning of the Long Count, but this hypothesis is questioned by some scholars. The identification of certain architectural assemblages as observatories of solstices and equinoxes is equally difficult to substantiate. So far, it has not been demonstrated how the Maya reckoned the seasons of their agricultural cycle or whether they observed the tropical or the sidereal year.

In colonial times, the star group known as the Pleiades was used to mark divisions of the night, and the constellation Gemini was also observed. A computation table in the *Dresden Codex* records intervals of possible eclipses of the Sun and Moon. Another correlates five revolutions of the planet Venus around the Sun with eight 365-day years and projects the count for 104 years, when it returns to the beginning Tzolkin date. Three sets of month positions associated with the cycle suggest its periodic correction. Other computations have not been adequately explained, among them some very long numbers that transcend the Long Count. Such numbers appear also on monuments and indicate a grandiose conception of the complexity and the almost infinite extent of time.

The Mexican (Aztec) calendar. The calendar of the Aztec was derived from earlier calendars in the Valley of Mexico and was basically similar to that of the Maya. The ritual day cycle was called Tonalpohualli and was formed, as was the Mayan Tzolkin, by the concurrence of a cycle of numerals 1 through 13 with a cycle of 20 day names, many of them similar to the day names of the Maya. The Tonalpohualli could be divided into four or five equal parts, each of four assigned to a world quarter and a colour and including the centre of the world if the parts were five. To the Aztec, the 13-day period defined by the day numerals was of prime importance, and each of 20 such periods was under the patronage of a specific deity. A similar list of 20 deities was associated with individual day names, and, in addition, there was a list of 13 deities designated as Lords of the Day, each accompanied by a flying creature, and a list of nine deities known as Lords of the Night. The lists of deities vary somewhat in different sources. They were probably used to determine the fate of the days by the Tonalpouhque, who were priests trained in calendrical divination. These priests were consulted as to lucky days whenever an important enterprise was undertaken or when a child was born. Children were often named after the day of their birth; and tribal gods, who were legendary heroes of the past, also bore calendar names.

The Aztec and Mayan systems compared

The Aztec year of 365 days was also similar to the year of the Maya, though probably not synchronous with it. It had 18 named months of 20 days each and an additional five days, called *nemontemi,* which were considered to be very unlucky. Though some colonial historians mention the use of intercalary days, in Aztec annals there is no indication of a correction in the length of the year. The years were named after days that fall at intervals of 365 days, and most scholars believe that these days held a fixed position in the year, though there appears to be some disagreement as to whether this position was the first day, the last day of the first month, or the last day of the last month, as was suggested by a distinguished Mexican scholar. Since 20 and 365 are both divisible by five, only four day names—Acatl (Reed), Tecpatl (Flint), Calli (House), and Tochtli (Rabbit)—figure in the names of the 52 years that form a cycle with the Tonalpohualli. The cycle begins with a year 2 Reed and ends with a year 1 Rabbit, which was regarded as a dangerous year of bad omen. At the end of such a cycle, all household utensils and idols were discarded and replaced by new ones, temples were renovated, and human sacrifice was offered to the Sun at midnight on a mountaintop as people awaited a new dawn.

The year served to fix the time of festivals, which took place at the end of each month. The new year was celebrated by the making of a new fire, and a more elaborate ceremony was held every four years, when the cycle had run through the four day names. Every eight years was celebrated the coincidence of the year with the 584-day period of the planet Venus, and two 52-year cycles formed "One Old Age," when the day cycle, the year, and the period of Venus all came together. All these periods were noted also by the Maya.

Where the Aztec differed most significantly from the Maya was in their more primitive number system and in their less precise way of recording dates. Normally, they noted only the day on which an event occurred and the name of the current year. This is ambiguous, since the same day, as designated in the way mentioned above, can occur twice in a year. Moreover, years of the same name recur at 52-year intervals, and Spanish colonial annals often disagree as to the length of time between two events. Other discrepancies in the records are only partially explained by the fact that different towns started their year with different months. The most widely accepted correlation of the calendar of Tenochtitlán with the Christian Julian calendar is based on the entrance of Cortés into that city on November 8, 1519, and on the surrender of Cuauhtémoc on August 13, 1521. According to this correlation, the first date was a day 8 Wind, the ninth day of the month Quecholli, in a year 1 Reed, the 13th year of a cycle.

Aztec "Calendar Stone." The calendar, discovered in 1790, is a basaltic monolith. It weighs approximately 25 tons and is about 3.7 metres in diameter. In the Museo Nacional de Antropología, Mexico City.

By courtesy of the Museo Nacional de Antropologia, Mexico City; photograph, Mexican National Tourist Council

The Mexicans, as all other Meso-Americans, believed in the periodic destruction and re-creation of the world. The "Calendar Stone" in the Museo Nacional de Antropología (National Museum of Anthropology) in Mexico City depicts in its central panel the date 4 Ollin (movement), on which they anticipated that their current world would be destroyed by earthquake, and within it the dates of previous holocausts: 4 Tiger, 4 Wind, 4 Rain, and 4 Water.

Peru: the Inca calendar. So little is known about the calendar used by the Inca that one can hardly make a statement about it for which a contrary opinion cannot be found. Some workers in the field even assert that there was no formal calendar but only a simple count of lunations. Since no written language was used by the Inca, it is impossible to check contradictory statements made by early colonial chroniclers. It is widely believed that the quipus of the Inca contain calendrical notations, but no satisfactory demonstration of this is possible.

Most historians agree that the Inca had a calendar based on the observation of both the Sun and the Moon, and their relationship to the stars. Names of 12 lunar months are recorded, as well as their association with festivities of the agricultural cycle; but there is no suggestion of the widespread use of a numerical system for counting time, although a quinary decimal system, with names of numbers at least up to 10,000, was used for other purposes. The organization of work on the basis of six weeks of nine days suggests the further possibility of a count by triads that could result in a formal month of 30 days.

The Chibcha system

A count of this sort was described by Alexander von Humboldt for a Chibcha tribe living outside of the Inca Empire, in the mountainous region of Colombia. The description is based on an earlier manuscript by a village priest, and one authority has dismissed it as "wholly imaginary," but this is not necessarily the case. The smallest unit of this calendar was a numerical count of three days, which, interacting with a similar count of 10 days, formed a standard 30-day "month." Every third year was made up of 13 moons, the others having 12. This formed a cycle of 37 moons, and 20 of these cycles made up a period of 60 years, which was subdivided into four parts and could be multiplied by 100. A period of 20 months is also mentioned. Although the account of the Chibcha system cannot be accepted at face value, if there is any truth in it at all it is suggestive of devices that may have been used also by the Inca.

In one account, it is said that the Inca Viracocha established a year of 12 months, each beginning with the New Moon, and that his successor, Pachacuti, finding confusion in regard to the year, built the sun towers in order to keep a check on the calendar. Since Pachacuti reigned less than a century before the conquest, it may be that the contradictions and the meagreness of information on the Inca calendar are due to the fact that the system was still in the process of being revised when the Spaniards first arrived.

(T.P.)

Despite the uncertainties, further research has made it clear that at least at Cuzco, the capital city of the Inca, there was an official calendar of the sidereal–lunar type, based on the sidereal month of 27⅓ days. It consisted of 328 nights (12 × 27⅓) and began on June 8/9, coinciding with the heliacal rising (the rising just after sunset) of the Pleiades; it ended on the first Full Moon after the June solstice (the winter solstice for the Southern Hemisphere). This sidereal–lunar calendar fell short of the solar year by 37 days, which consequently were intercalated. This intercalation, and thus the place of the sidereal–lunar within the solar year, was fixed by following the cycle of the Sun as it "strengthened" to summer (December) solstice and "weakened" afterward, and by noting a similar cycle in the visibility of the Pleiades.

(T.P./C.A.R.)

North American Indian time counts. No North American Indian tribe had a true calendar—a single integrated system of denoting days and longer periods of time. Usually, intervals of time were counted independently of one another. The day was a basic unit recognized by all tribes, but there is no record of aboriginal names for days. A common device for keeping track of days was a bundle of sticks of known number, from which one was extracted for every day that passed, until the bundle was exhausted. Longer periods of time were usually counted by moons, which began with the New Moon, or conjunction of the Sun and Moon. Years were divided into four seasons, occasionally five, and when counted were usually designated by one of the seasons; *e.g.,* a North American Indian might say that a certain event had happened 10 winters ago. Among sedentary agricultural tribes, the cycle of the seasons was of great ritual importance, but the time of the beginning of the year varied. Some observed it about the time of the vernal equinox, others in the fall. The Hopi tribe of northern Arizona held a new-fire ceremony in November. The Creek ceremony, known as the "Busk," was held late in July or in August, but it is said that each Creek town or settlement set its own date for the celebration.

As years were determined by seasons and not by a fixed number of days, the correlation of moons and years was also approximate and not a function of a daily count. Most tribes reckoned 12 moons to a year. Some northern tribes, notably those of New England, and the Cree tribes, counted 13. The Indians of the northwest coast divided their years into two parts, counting six moons to each part, and the Kiowa split one of their 12 moons between two unequal seasons, beginning their year with a Full Moon.

The naming of moons

The naming of moons is perhaps the first step in transforming them into months. The Zuni Indians of New Mexico named the first six moons of the year, referring to the remainder by colour designations associated with the four cardinal (horizontal) directions, and the zenith and the nadir. Only a few Indian tribes attempted a more precise correlation of moons and years. The Creeks are said to have added a moon between each pair of years, and the Haida from time to time inserted a "between moon" in the division of their year into two parts. It is said that an unspecified tribe of the Sioux or the Ojibwa (Chippewa) made a practice of adding a "lost moon" when 30 moons had waned.

A tally of years following an important event was sometimes kept on a notched stick. The best known record commemorates the spectacular meteoric shower (the Leonids) of 1833. Some northern tribes recorded series of events by pictographs, and one such record, said to have been originally painted on a buffalo robe and known as the "Lone-dog Winter Count," covers a period of 71 years beginning with 1800.

Early explorers had little opportunity to learn about the calendrical devices of the Indians, which were probably held sacred and secret. Contact with Europeans and their Christian calendar doubtless altered many aboriginal practices. Thus, present knowledge of the systems used in the past may not reflect their true complexity.

(T.P.)

The Western calendar and calendar reforms

The calendar now in general worldwide use had its origin in the desire for a solar calendar that kept in step with the seasons and possessed fixed rules of intercalation. Because it developed in Western Christendom, it had also to provide a method for dating movable religious feasts, the timing of which had been based on a lunar reckoning. To reconcile the lunar and solar schemes, features of the Roman republican calendar and the Egyptian calendar were combined.

The Roman republican calendar was basically a lunar reckoning and became increasingly out of phase with the seasons as time passed. By about 50 BC the vernal equinox that should have fallen late in March fell on the Ides of May, some eight weeks later, and it was plain that this error would continue to increase. Moreover, the behaviour of the Pontifices (see above *The early Roman calendar*) made it necessary to seek a fixed rule of intercalation in order to put an end to arbitrariness in inserting months.

In addition to the problem of intercalation, it was clear that the average Roman republican year of 366.25 days would always show a continually increasing disparity with the seasons, amounting to one month every 30 years, or three months a century. But the great difficulty facing any reformer was that there seemed to be no way of effecting a change that would still allow the months to remain in step with the phases of the Moon and the year with the seasons. It was necessary to make a fundamental break with traditional reckoning to devise an efficient seasonal calendar.

THE JULIAN CALENDAR

In the mid-1st century BC Julius Caesar invited Sosigenes, an Alexandrian astronomer, to advise him about the reform of the calendar, and Sosigenes decided that the only practical step was to abandon the lunar calendar altogether. Months must be arranged on a seasonal basis, and a tropical (solar) year used, as in the Egyptian calendar, but with its length taken as 365¼ days.

To remove the immense discrepancy between calendar date and equinox, it was decided that the year known in modern times as 46 BC should have two intercalations. The first was the customary intercalation of the Roman republican calendar due that year, the insertion of 23 days following February 23. The second intercalation, to bring the calendar in step with the equinoxes, was achieved by inserting two additional months between the end of November and the beginning of December. This insertion amounted to an addition of 67 days, making a year of no less than 445 days and causing the beginning of March, 45 BC in the Roman republican calendar, to fall on what is still called January 1 of the Julian calendar.

Previous errors having been corrected, the next step was to prevent their recurrence. Here Sosigenes' suggestion about a tropical year was adopted and any pretense to

Leap year

a lunar calendar was rejected. The figure of 365.25 days was accepted for the tropical year, and, to achieve this by a simple civil reckoning, Caesar directed that a calendar year of 365 days be adopted and that an extra day be intercalated between February 23 and 24 every fourth year. Since February ordinarily had 28 days, February 23 was the sixth day before the Kalendae, or beginning of March, and was known as the *sexto-kalendae;* the intercalary day, when it appeared, came the day after and was therefore called the *bis-sexto-kalendae.* This practice led to the term bissextile being used to refer to such a leap year. The name leap year is a later connotation, probably derived from the Old Norse *hlaupa* ("to leap") and used because, in a bissextile year, any fixed festival after February leaps forward, falling on the second weekday from that on which it fell the previous year, not on the next weekday as it would do in an ordinary year.

In Caesar's edict, the intercalary day was known as a *punctum temporis* (point of time), and anyone born that day had subsequent birthdays on February 23 (the day before); but lawyers then and in medieval times raised a number of arguments about its precise interpretation. Also, the Pontifices misinterpreted the edict and inserted the intercalation too frequently. The error arose because of the Roman practice of inclusive numbering, so that an intercalation once every fourth year meant to them intercalating every three years, because a bissextile year was counted as the first year of the next four-year period. This error continued undetected for 36 years, during which period 12 days instead of nine were added. The emperor Augustus made a correction by omitting intercalary days between 8 BC and AD 8. In consequence it was not until several decades after its inception that the Julian calendar came into proper operation, a fact that is important in chronology but is often forgotten.

Months of the Julian calendar

It seems that the months of the Julian calendar were taken over from the Roman republican calendar but were slightly modified to give a more even pattern of numbering. The republican calendar months of March, May, and Quintilis (July), which had each possessed 31 days, were retained unaltered. Although there is doubt about the details, changes may have occurred as follows. Except for October, all the months that had previously had only 29 days had either one or two days added. January, September, and November received two days, bringing their totals to 31, while April, June, Sextilis (August), and December received one day each, bringing their totals to 30. October was reduced by one day to a total of 30 days and February increased to 29 days, or 30 in a bissextile year. With the exception of February, the scheme resulted in months having 30 or 31 days alternately throughout the year. And in order to help farmers, Caesar issued an almanac showing on which dates of his new calendar various seasonal astronomical phenomena would occur.

These arrangements for the months can only have remained in force for a short time, because in 8 BC changes were made by Augustus. In 44 BC, the second year of the Julian calendar, the Senate proposed that the name of the month Quintilis be changed to Julius (July), in honour of Julius Caesar, and in 8 BC the name of Sextilis was similarly changed to Augustus (August). Perhaps because Augustus felt that his month must have at least as many days as Julius Caesar's, February was reduced to 28 days and August increased to 31. But because this made three 31-day months (July, August, and September) appear in succession, Augustus is supposed to have reduced September to 30 days, added a day to October to make it 31 days, reduced November by one day to 30 days, and increased December from 30 to 31 days, giving the months the lengths they have today.

Several scholars, however, believe that Caesar originally left February with 28 days (to avoid affecting certain religious rites in honour of the gods of the netherworld) and added two days to Sextilis for a total of 31; January, March, May, Quintilis, October, and December also had 31 days, with 30 days for April, June, September, and November. The subsequent change of Sextilis to Augustus thus involved no addition of days to the latter.

Divisions of the month

The Julian calendar retained the Roman republican calendar method of numbering the days of the month. Compared with the present system, the Roman numbering seems to run backward, for the first day of the month was known as the Kalendae, but subsequent days were not enumerated as so many after the Kalendae but as so many before the following Nonae ("nones"), the day called nonae being the ninth day before the Ides (from *iduare,* meaning "to divide"), which occurred in the middle of the month and were supposed to coincide with the Full Moon. Days after the Nonae and before the Ides were numbered as so many before the Ides, and those after the Ides as so many before the Kalendae of the next month.

There were no weeks in the original Julian calendar, but days were designated either *dies fasti* or *dies nefasti,* the former being business days and days on which the courts were open; this had been the practice in the Roman republican calendar. Julius Caesar designated his additional days all as *dies fasti,* and they were added at the end of the month so that there was no interference with the dates traditionally fixed for *dies comitiales* (days when public assemblies might be convened) and *dies festi* and *dies feriae* (religious festivals and holy days). Originally, then, the Julian calendar had a permanent set of dates for administrative matters. The official introduction of the seven-day week by Emperor Constantine I in the 4th century AD disrupted this arrangement.

Fixing of January as the first month

It appears, from the date of insertion of the intercalary month in the Roman republican calendar and the habit of designating years by the names of the consuls, that the calendar year had originally commenced in March, which was the date when the new consul took office. In 222 BC the date of assuming duties was fixed as March 15, but in 153 BC it was transferred to the Kalendae of January, and there it remained. January therefore became the first month of the year, and in the western region of the Roman Empire, this practice was carried over into the Julian calendar. In the eastern provinces, however, years were often reckoned from the accession of the reigning emperor, the second beginning on the first New Year's day after the accession; and the date on which this occurred varied from one province to another.

THE GREGORIAN CALENDAR

The Julian calendar year of 365.25 days was too long, since the correct value for the tropical year is 365.242199 days. This error of 11 minutes 14 seconds per year amounted to almost one and a half days in two centuries, and seven days in 1,000 years. Once again the calendar became increasingly out of phase with the seasons. From time to time, the problem was placed before church councils, but no action was taken because the astronomers who were consulted doubted whether enough precise information was available for a really accurate value of the tropical year to be obtained. By 1545, however, the vernal equinox, which was used in determining Easter, had moved 10 days from its proper date; and in December, when the Council of Trent met for the first of its sessions, it authorized Pope Paul III to take action to correct the error. Correction required a solution, however, that neither Paul III nor his successors were able to obtain in satisfactory form until nearly 1572, the year of election of Pope Gregory XIII. Gregory found various proposals awaiting him and agreed to issue a bull that the Jesuit astronomer Christopher Clavius (1537–1612) began to draw up, using suggestions made by the astronomer and physician Luigi Lilio (also known as Aloysius Lilius; died 1576).

The papal bull appeared in February 1582. First, in order to bring the vernal equinox back to March 21, the day following the Feast of St. Francis (that is, October 5) was to become October 15, thus omitting 10 days. Second, to bring the year closer to the true tropical year, a value of 365.2422 days was accepted. This value differed by 0.0078 days per year from the Julian calendar reckoning, amounting to 0.78 days per century, or 3.12 days every 400 years. It was therefore promulgated that three out of every four centennial years should be common years, that is, not leap years; and this practice led to the rule that no centennial years should be leap years unless exactly divisible by 400. Thus, 1700, 1800, and 1900 were not leap years, as they would have been in the Julian calendar, but the year 2000 will be. The bull also laid down rules for calculating the date of Easter.

The date of Easter; epacts. Easter was the most important feast of the Christian Church, and its place in the calendar determined the position of the rest of the church's movable feasts (see CHRISTIANITY: *Church year*). Because its timing depended on both the Moon's phases and the vernal equinox, ecclesiastical authorities had to seek some way of reconciling lunar and solar calendars. Some simple form of computation, usable by non-astronomers in remote places, was desirable. There was no easy or obvious solution, and to make things more difficult there was no unanimous agreement on the way in which Easter should be calculated, even on a lunar calendar.

Dependence on the Passover

Easter, being the festival of the Resurrection, had to depend on the dating of the Crucifixion, which occurred three days earlier and just before the Jewish Passover. The Passover was celebrated on the 14th day of Nisan, the first month in the Jewish religious year—that is, the lunar month the 14th day of which falls on or next after the vernal equinox. The Christian churches in the eastern Mediterranean area celebrated Easter on the 14th of Nisan on whatever day of the week it might fall, but the rest of Christendom adopted a more elaborate reckoning to ensure that it was celebrated on a Sunday in the Passover week.

To determine precisely how the Resurrection and Easter Day should be dated, reference was made to the Gospels; but even as early as the 2nd century AD, difficulties had arisen since the synoptic Gospels (Matthew, Mark, and Luke) appeared to give a different date from the Gospel According to John for the Crucifixion. This difference led to controversy that was later exacerbated by another difficulty caused by the Jewish reckoning of a day from sunset to sunset. The question arose of how the evening of the 14th day should be calculated, and some—the Quintodecimans—claimed that it meant one particular evening, but others—the Quartodecimans—claimed that it meant the evening before, since sunset heralded a new day. Both sides had their protagonists, the Eastern churches supporting the Quartodecimans, the Western churches the Quintodecimans. The question was finally decided at the Council of Nicaea, in 325, in favour of the Quintodecimans, and the Western church agreed. The Eastern churches decided to retain the Quartodeciman position, and the church in England, which had few links with the European churches at this time, retained the Quartodeciman position until Roman missionaries arrived in the 6th century, when a change was made. The dating of Easter in the Gregorian calendar was based on the decision of the Council of Nicaea, which decreed that Easter should be celebrated on the Sunday immediately following the (Paschal) Full Moon that fell on or after the vernal equinox, which they took as March 21. The council also ordered that if this Full Moon fell on a Sunday, the festival should be held seven days later.

Dominical letters

With these provisions in mind, the problem could be broken down into two parts: first, devising a simple but effective way of calculating the days of the week for any date in the year and, second, determining the date of the Full Moons in any year. The first part was solved by the use of a letter code derived from a similar Roman system adopted for determining market days. For ecclesiastical use, the code gave what was known as the Sunday, or dominical, letter.

The seven letters A through G are each assigned to a day, consecutively from January 1 so that January 1 appears as A, January 2 as B, to January 7 which appears as G, the cycle then continuing with January 8 as A, January 9 as B, and so on. Then in any year the first Sunday is bound to be assigned to one of the letters A–G in the first cycle, and all Sundays in the year possess that dominical letter. For example, if the first Sunday falls on January 3, C will be the dominical letter for the whole year. No dominical letter is placed against the intercalary day, February 29, but since it is still counted as a weekday and given a name, the series of letters moves back one day every leap year after intercalation. Thus, a leap year beginning with the dominical letter C will change to a year with the dominical letter D on March 1; and in lists of dominical letters, all leap years are given a double letter notation, in the example just quoted, CD. It is not difficult to see what dominical letter or letters apply to any particular year, and it is also a comparatively simple matter to draw up a table of dominical letters for use in determining Easter Sunday. The possible dates on which Easter Sunday can fall are written down—they run from March 22 through April 25—and against them the dominical letters for a cycle of seven years. Once the dominical letter for a year is known, the possible Sundays for celebrating Easter can be read directly from the table. This system does not, of course, completely determine Easter; to do so, additional information is required.

Golden numbers

This must provide dates for Full Moons throughout the year, and for this a lunar cycle like the Metonic cycle was originally used. Tables were prepared, again using the range of dates on which Easter Sunday could appear, and against each date a number from one through to 19 was placed. This number indicated which of the 19 years of the lunar cycle would give a Full Moon on that day. From medieval times these were known as golden numbers, possibly from a name used by the Greeks for the numbers on the Metonic cycle or because gold is the colour used for them in manuscript calendars.

The system of golden numbers was introduced in 530, but the numbers were arranged as they should have been if adopted at the Council of Nicaea two centuries earlier; and the cycle was taken to begin in a year when the New Moon fell on January 1. Working backward, chronologers found that this date had occurred in the year preceding AD 1, and therefore the golden number for any year is found by dividing the year by 19, then adding one to the remainder. If the result is zero, the golden number for the year is 19.

To compute the date of Easter, the medieval chronologer computed the golden number for the year and then consulted his table to see by which date this number lay. Having found this date, that of the first Full Moon after March 20, he consulted his table of dominical letters and saw the next date against which the dominical letter for that year appeared; this was the Sunday to be designated Easter. The method, modified for dropping centennial leap years as practiced in the Gregorian calendar, is still given in the English prayer book, although it was officially discarded when the Gregorian calendar was introduced.

The system of golden numbers was eventually rejected because the astronomical Full Moon could differ by as much as two days from the date they indicated. It was Lilius who had proposed a more accurate system based on one that had already been in use unofficially while the Julian calendar was still in force. Called the epact—the word is derived from the Greek *epagein,* meaning "to intercalate"—this was again a system of numbers concerned with the Moon's phases, but now indicating the age of the Moon on the first day of the year, and from which the age of the Moon on any day of the year may be found, at least approximately, by counting, using alternately months of 29 and 30 days.

Lunar and solar corrections

The epact as previously used was not, however, completely accurate because, like the golden number, it had been based on the Metonic cycle. This 19-year cycle was in error, the discrepancy amounting to eight days every 2,500 years. A one-day change on certain centennial years was then instituted by making the computed age of the Moon one day later seven times, at 300-year intervals, and an eighth time after a subsequent 400 years. This operation was known as the lunar correction, but it was not the only correction required; there was another.

Because the Gregorian calendar used a more accurate value for the tropical year than the Julian calendar and achieved this by omitting most centennial leap years, Clavius decided that when the cycle of epacts reached an ordinary centennial year, the number of the epact should be reduced by one; this reduction became known as the solar correction.

One advantage of the epact number was that it showed the age of the Moon on January 1 and so permitted a simple calculation of the dates of New Moon and Full Moon for the ensuing year. Another was that it lent itself to the construction of cycles of 30 epact numbers, each diminishing by one from the previous cycle, so that when it became necessary at certain centennial years to shift from one cycle to another, there would still be a cycle ready that retained a correct relationship between dates and New Moons.

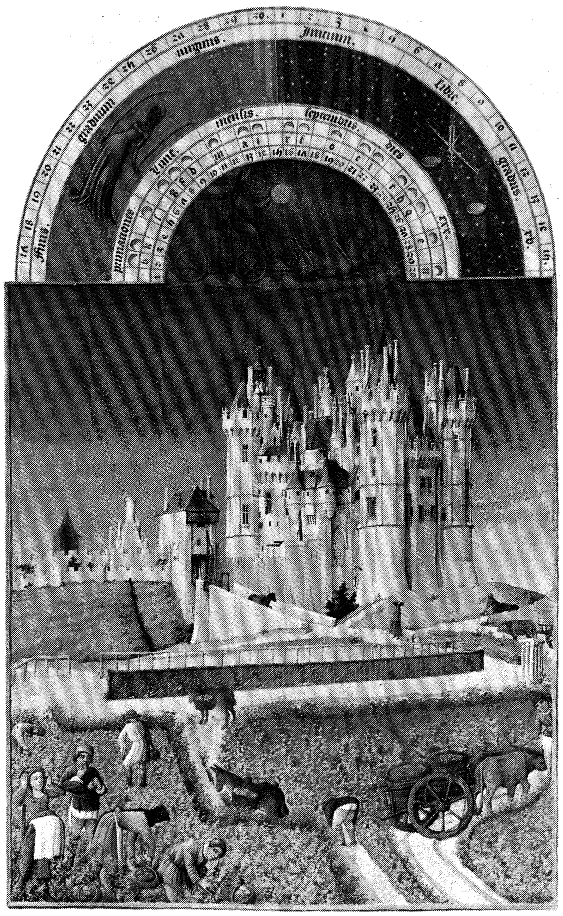
For determining Easter, a table was prepared of the golden numbers, one through 19, and below them the cycles of epacts for about 7,000 years; after this time, all the epact cycles are repeated. A second table was then drawn up, giving the dates of Easter Full Moons for different epact numbers. Once the epact for the year was known, the date of the Easter Full Moon could be immediately obtained, while consultation of a table of dominical letters showed which was the next Sunday. Thus, the Gregorian system of epacts, while more accurate than the old golden numbers, still forced the chronologer to consult complex astronomical tables.

Adoption in various countries. The derivation of the term style for a type of calendar seems to have originated some time soon after the 6th century as a result of developments in calendar computation in the previous 200 years. In AD 463, Victorius (or Victorinus) of Aquitaine, who had been appointed by Pope Hilarius to undertake calendar revision, devised the Great Paschal (*i.e.,* Passover) period, sometimes later referred to as the Victorian period. It was a combination of the solar cycle of 28 years and the Metonic 19-year cycle, bringing the Full Moon back to the same day of the month, and amounted to 28 × 19, or 532 years. In the 6th century, this period was used by Dionysius Exiguus (Denis the Little) in computing the date of Easter, because it gave the day of the week for any day in any year, and so it also became known as the Dionysian period. Dionysius took the year now called AD 532 as the first year of a new Great Paschal period and the year now designated 1 BC as the beginning of the previous cycle. In the 6th century it was the general belief that this was the year of Christ's birth, and because of this Dionysius introduced the concept of numbering years consecutively through the Christian Era. The method was adopted by some scholars but seems only to have become widely used after its popularization by the Venerable Bede of Jarrow (673?–735), whose reputation for scholarship was very high in Western Christendom in the 8th century. This system of BC/AD numbering threw into relief the different practices, or styles, of reckoning the beginning of the year then in use. When the Gregorian calendar firmly established January 1 as the beginning of its year, it was widely referred to as the New Style calendar, with the Julian the Old Style calendar. In Britain, under the Julian calendar, the year had first begun on December 25 and then, from the 14th century onward, on March 25.

The Great Paschal period

Because of the division of the Eastern and Western Christian churches and of Protestants and Roman Catholics, the obvious advantages of the Gregorian calendar were not accepted everywhere, and in some places adoption was extremely slow. In France, Italy, Luxembourg, Portugal, and Spain, the New Style calendar was adopted in 1582, and it was in use by most of the German Roman Catholic states as well as by Belgium and part of the Netherlands by 1584. Switzerland's change was gradual, on the other hand, beginning in 1583 and being completed only in 1812. Hungary adopted the New Style in 1587, and then there was a pause of more than a century before the first Protestant countries made the transition from the Old Style calendar. In 1699–1700, Denmark and the Dutch and German Protestant states embraced the New Style, although the Germans declined to adopt the rules laid down for determining Easter. The Germans preferred to rely instead on astronomical tables and specified the use of the *Tabulae Rudolphinae* (*Rudolphine Tables*), based on the 16th-century observations of Tycho Brahe. They acceded to the Gregorian calendar rules for Easter only in 1776. Britain adopted the New Style in 1752 and Sweden in 1753, although the Swedes, because they had in 1740 followed the German Protestants in using their astronomical methods for determining Easter, declined to adopt the Gregorian calendar rules until 1844. Japan adopted the New Style in 1873; Egypt adopted it in 1875; and between 1912 and 1917 it was accepted by Albania, Bulgaria, China, Estonia, Latvia, Lithuania, Romania, Turkey, and Yugoslavia. The Soviet Union adopted the New Style in 1918; Greece in 1923.

Adoption of the New Style by Protestant countries



Calendar illustration for September from the *Très Riches Heures du duc de Berry,* manuscript illuminated by the Limburg brothers, 1416. In the Musée Condé, Chantilly, France.

By courtesy of the Musee Conde, Chantilly; photograph, Giraudon—Art Resource

In Britain and the British dominions, the change was made when the difference between the New and Old Style calendars amounted to 11 days: the lag was covered by naming the day after September 2, 1752, as September 14, 1752. There was widespread misunderstanding among the public, however, even though legislation authorizing the change had been framed to avoid injustice and financial hardship. The Alaskan territory retained the Old Style calendar until 1867, when it was transferred from Russia to the United States.

CALENDAR REFORM SINCE THE MID-18TH CENTURY

The French republican calendar. In late 18th-century France, with the approach of the French Revolution, demands began to be made for a radical change in the civil calendar that would divorce it completely from any ecclesiastical connections. The first attacks on the Gregorian calendar and proposals for reform came in 1785 and 1788, the changes being primarily designed to divest the calendar of all its Christian associations. After the storming of the Bastille in July 1789, demands became more vociferous, and a new calendar, to start from "the first year of liberty," was widely spoken about. In 1793 the National Convention appointed Charles-Gilbert Romme, president of the committee of public instruction, to take charge of the reform. Technical matters were entrusted to the mathematicians Joseph-Louis Lagrange and Gaspard Monge and the renaming of the months to the Paris deputy to the convention, Philippe Fabre d'Églantine. The results of their deliberations were submitted to the convention in September of the same year and were immediately accepted, it being promulgated that the new calendar should become law on October 5.

The French republican calendar, as the reformed system came to be known, was taken to have begun on September 22, 1792, the day of the proclamation of the Republic and, in that year, the date also of the autumnal equinox. The total number of days in the year was fixed at 365, the same as in the Julian and Gregorian calendars, and this was divided into 12 months of 30 days each, the remaining five days at year's end being devoted to festivals and vacations. These were to fall between September 17 and 22 and were specified, in order, to be festivals in honour of virtue, genius, labour, opinion, and rewards. In a leap year an extra festival was to be added—the festival of the Revolution. Leap years were retained at the same frequency as in the Gregorian calendar, but it was enacted that the first leap year should be year 3, not year 4 as it would have been if the Gregorian calendar had been followed precisely in this respect. Each four-year period was to be known as a *Franciade.*

Weeks and *décades*

The seven-day week was abandoned, and each 30-day month was divided into three periods of 10 days called *décades,* the last day of a *décade* being a rest day. It was also agreed that each day should be divided into decimal parts, but this was not popular in practice and was allowed to fall into disuse.

The months themselves were renamed so that all previous associations should be lost, and Fabre d'Églantine chose descriptive names as follows (the descriptive nature and corresponding Gregorian calendar dates for years 1, 2, 3, 5, 6, and 7 are given in parentheses): Vendémiaire ("vintage," September 22 to October 21), Brumaire ("mist," October 22 to November 20), Frimaire ("frost," November 21 to December 20), Nivôse ("snow," December 21 to January 19), Pluviôse ("rain," January 20 to February 18), Ventôse ("wind," February 19 to March 20), Germinal ("seedtime," March 21 to April 19), Floréal ("blossom," April 20 to May 19), Prairial ("meadow," May 20 to June 18), Messidor ("harvest," June 19 to July 18), Thermidor ("heat," July 19 to August 17), and Fructidor ("fruits," August 18 to September 16).

The French republican calendar was short-lived, for while it was satisfactory enough internally, it clearly made for difficulties in communication abroad because its months continually changed their relationship to dates in the Gregorian calendar. In September 1805, under the Napoleonic regime, the calendar was virtually abandoned, and on January 1, 1806, it was replaced by the Gregorian calendar.

Soviet calendar reforms. When Soviet Russia undertook its calendar reform in February 1918, it merely moved from the Julian calendar to the Gregorian. This move resulted in a loss of 13 days, so that February 1, 1918, became February 14.

Modern schemes for reform. The current calendar is not without defects, and reforms are still being proposed. Astronomically, it really calls for no improvement, but the seven-day week and the different lengths of months are unsatisfactory to some. Clearly, if the calendar could have all festivals and all rest days fixed on the same dates every year, as in the original Julian calendar, this arrangement would be more convenient, and two general schemes have been put forward—the International Fixed Calendar and the World Calendar.

The International Fixed Calendar is essentially a perpetual Gregorian calendar, in which the year is divided into 13 months, each of 28 days, with an additional day at the end. Present month names are retained, but a new month named Sol is intercalated between June and July. The additional day follows December 28 and bears no designation of month date or weekday name, while the same would be true of the day intercalated in a leap year after June 28. In this calendar, every month begins on a Sunday and ends on a Saturday.

It is claimed that the proposed International Fixed Calendar does not conveniently divide into quarters for business reckoning; and the World Calendar is designed to remedy this deficiency, being divided into four quarters of 91 days each, with an additional day at the end of the year. In each quarter, the first month is of 31 days and the second and third of 30 days each. The extra day comes after December 30 and bears no month or weekday designation, nor does the intercalated leap year day that follows June 30. In the World Calendar January 1, April 1, July 1, and October 1 are all Sundays. Critics point out that each month extends over part of five weeks, and each month within a given quarter begins on a different day. Nevertheless, both these proposed reforms seem to be improvements over the present system that contains so many variables. (C.A.R.)

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(C.A.R./J.D.Sc./E.J.Bi./E.J.Wi./N.A.Z./J.A.B.v.B./Ch.L./T.P.)

Calvin and Calvinism

John Calvin, in French Jean Calvin or Cauvin (1509–1564), was the leading French Protestant Reformer and the most important figure in the second generation of the Protestant Reformation. His interpretation of Christianity, advanced above all in his *Institutio Christianae religionis* (1536 but elaborated in later editions; *Institutes of the Christian Religion*), and the institutional and social patterns he worked out for Geneva deeply influenced Protestantism elsewhere in Europe and in North America. The Calvinist form of Protestantism is widely thought to have had a major impact on the formation of the modern world.

This article deals with the man, his achievements, and the Calvinist tradition. For a further treatment of Calvinism, see PROTESTANTISM.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 827 and 961.

This article is divided into the following sections:

CALVIN

Life and works. Calvin was born on July 10, 1509, in Noyon, in Picardy, France, the son of middle-class parents. His father, a lay administrator in the service of the local bishop, sent him to the University of Paris in 1523 to be educated for the priesthood but later decided that he should be a lawyer; from 1528 to 1531, therefore, Calvin studied in the law schools of Orléans and Bourges. He then returned to Paris. During these years he was also exposed to Renaissance humanism, influenced by Erasmus and Jacques Lefèvre d'Étaples, which constituted the radical student movement of the time. This movement, which antedates the Reformation, aimed to reform church and society on the model of both classical and Christian antiquity, to be established by a return to the Bible studied in its original languages. It left an indelible mark on Calvin. Under its influence he studied Greek and Hebrew as well as Latin, the three languages of ancient Christian discourse, in preparation for serious study of the Scriptures. It also intensified his interest in the classics; his first publication (1532) was a commentary on Seneca's essay on clemency. But the movement, above all, emphasized salvation of individuals by grace rather than good works and ceremonies.

The influence of Renaissance humanism

Flight to Switzerland

Calvin's Paris years came to an abrupt end late in 1533. Because the government became less tolerant of this reform movement, Calvin, who had collaborated in the preparation of a strong statement of theological principles for a public address delivered by Nicolas Cop, rector of the university, found it prudent to leave Paris. Eventually he made his way to Basel, then Protestant but tolerant of religious variety. Up to that point, however, there is little evidence of Calvin's conversion to Protestantism, an event difficult to date because it was probably gradual. His beliefs before his flight to Switzerland were probably not incompatible with Roman Catholic orthodoxy. But they underwent a change when he began to study theology intensively in Basel. Probably in part to clarify his own beliefs, he began to write. He began with a preface to a French translation of the Bible by his cousin Pierre Olivétan and then undertook what became the first edition of the *Institutes,* his masterwork, which, in its successive revisions, became the single most important statement of Protestant belief. Calvin published later editions in both Latin and French, containing elaborated and in a few cases revised teachings and replies to his critics. The final versions appeared in 1559 and 1560. The *Institutes* also reflected the findings of Calvin's massive biblical commentaries, which, presented extemporaneously in Latin as lectures to ministerial candidates from many countries, make up the largest proportion of his works. In addition he wrote many theological and polemical treatises.

Protestantism in Geneva

The 1536 *Institutes* had given Calvin some reputation among Protestant leaders. Therefore, on discovering that Calvin was spending a night in Geneva late in 1536, the Reformer and preacher Guillaume Farel, then struggling to plant Protestantism in that town, persuaded him to remain to help in this work. The Reformation was in trouble in Geneva, a town of about 10,000 where Protestantism had only the shallowest of roots. Other towns in the region, initially ruled by their prince-bishops, had successfully won self-government much earlier, but Geneva had lagged behind in this process largely because its prince-bishop was supported by the neighbouring duke of Savoy. There had been iconoclastic riots in Geneva in the mid-1520s, but these had negligible theological foundations. Protestantism had been imposed on religiously unawakened Geneva chiefly as the price of military aid from Protestant Bern. The limited enthusiasm of Geneva for Protestantism, reflected by a resistance to religious and moral reform, continued almost until Calvin's death. The resistance was all the more serious because the town council in Geneva, as in other Protestant towns, exercised ultimate control over the church and the ministers, all French refugees. The main issue was the right of excommunication, which the ministers regarded as essential to their authority but which the council refused to concede. The uncompromising attitudes of Calvin and Farel finally resulted in their expulsion from Geneva in May 1538.

By courtesy of the Museum Boymans-van Beuningen, Rotterdam

Calvin, oil painting by an anonymous master of the French school, *c.* 1550. In the Boymans-van Beuningen Museum, Rotterdam.

Calvin found refuge for the next three years in the German Protestant city of Strasbourg, where he was pastor of a church for French-speaking refugees and also lectured on the Bible; there he published his commentary on the Letter of Paul to the Romans. There too, in 1540, he married Idelette de Bure, the widow of a man he had converted from Anabaptism. Although none of their children survived infancy, their marital relationship proved to be extremely warm. During his Strasbourg years Calvin also learned much about the administration of an urban church from Martin Bucer, its chief pastor. Meanwhile Calvin's attendance at various international religious conferences made him acquainted with other Protestant leaders and gave him experience in debating with Roman Catholic theologians. Henceforth he was a major figure in international Protestantism.

Return to Geneva

In September 1541 Calvin was invited back to Geneva, where the Protestant revolution, without strong leadership, had become increasingly insecure. Because he was now in a much stronger position, the town council in November enacted his *Ecclesiastical Ordinances,* which provided for the religious education of the townspeople, especially children, and instituted Calvin's conception of church order. It also established four groups of church officers: pastors and teachers to preach and explain the Scriptures, elders representing the congregation to administer the church, and deacons to attend to its charitable responsibilities. In addition it set up a consistory of pastors and elders to make all aspects of Genevan life conform to God's law. It undertook a wide range of disciplinary actions covering everything from the abolition of Roman Catholic "superstition" to the enforcement of sexual morality, the regulation of taverns, and measures against dancing, gambling, and swearing. These measures were resented by a significant element of the population, and the arrival of increasing numbers of French religious refugees in Geneva was a further cause of native discontent. These tensions, as well as the persecution of Calvin's followers in France, help to explain the trial and burning of Michael Servetus, a Spanish theologian preaching and publishing unorthodox beliefs. When Servetus unexpectedly arrived in Geneva in 1553, both sides felt the need to demonstrate their zeal for orthodoxy. Calvin was responsible for Servetus' arrest and conviction, though he had preferred a less brutal form of execution.

The struggle over control of Geneva lasted until May 1555, when Calvin finally prevailed and could devote himself more wholeheartedly to other matters. He had constantly to watch the international scene and to keep his Protestant allies in a common front. Toward this end he engaged in a massive correspondence with political and religious leaders throughout Protestant Europe. He also continued his commentaries on Scripture, working through the whole New Testament except the Revelation to John and most of the Old Testament. Many of these commentaries were promptly published, often with dedications to such European rulers as Queen Elizabeth, though Calvin had too little time to do much of the editorial work himself. Committees of amanuenses took down what he said, prepared a master copy, and then presented it to Calvin for approval. During this period Calvin also established the Genevan Academy to train students in humanist learning in preparation for the ministry and positions of secular leadership. He also performed a wide range of pastoral duties, preaching regularly and often, doing numerous weddings and baptisms, and giving spiritual advice. Worn out by so many responsibilities and suffering from a multitude of ailments, he died on May 27, 1564.

Scriptural commentaries

Personality. Unlike Martin Luther, Calvin was a reticent man; he rarely expressed himself in the first person singular. This reticence has contributed to his reputation as cold, intellectual, and humanly unapproachable. His thought, from this perspective, has been interpreted as abstract and concerned with timeless issues rather than as the response of a sensitive human being to the needs of a particular historical situation. Those who knew him, however, perceived him differently, remarking on his talent for friendship but also on his hot temper. Moreover, the intensity of his grief on the death of his wife, as well as his empathic reading of many passages in Scripture, revealed a large capacity for feeling.

Calvin's facade of impersonality can now be understood as concealing an unusually high level of anxiety about the world around him, about the adequacy of his own efforts to deal with its needs, and about human salvation, notably including his own. He believed that every Christian—and he certainly included himself—suffers from terrible bouts of doubt. From this perspective the need for control both of oneself and the environment, often discerned in Calvinists, can be understood as a function of Calvin's own anxiety.

Calvin's anxiety found expression in two metaphors for the human condition that appear again and again in his writings: as an abyss in which human beings have lost their way and as a labyrinth from which they cannot escape. Calvinism as a body of thought must be understood as the product of Calvin's effort to escape from the terrors conveyed by these metaphors.

Intellectual formation. Historians are generally agreed that Calvin is to be understood primarily as a Renaissance humanist who aimed to apply the novelties of humanism to recover a biblical understanding of Christianity. Thus he sought to appeal rhetorically to the human heart rather than to compel agreement, in the traditional manner of systematic theologians, by demonstrating dogmatic truths. His chief enemies, indeed, were the systematic theologians of his own time, the Scholastics, both because they relied too much on human reason rather than the Bible and because their teachings were lifeless and irrelevant to a world in desperate need. Calvin's humanism meant first that he thought of himself as a biblical theologian in accordance with the Reformation slogan *scriptura sola.* He was prepared to follow Scripture even when it surpassed the limits of human understanding, trusting to the Holy Spirit to inspire faith in its promises. Like other humanists, he was also deeply concerned to remedy the evils of his own time; and here too he found guidance in Scripture. Its teachings could not be presented as a set of timeless abstractions but had to be brought to life by adapting them to the understanding of contemporaries according to the rhetorical principle of decorum—*i.e.,* suitability to time, place, and audience.

Calvin's humanism

Calvin's humanism influenced his thought in two other basic ways. For one, he shared with earlier Renaissance humanists an essentially biblical conception of the human personality, comprehending it not as a hierarchy of faculties ruled by reason but as a mysterious unity in which what is primary is not what is highest but what is central: the heart. This conception assigned more importance to will and feelings than to the intellect, and it also gave new dignity to the body. For this reason Calvin rejected the ascetic disregard of the body's needs that was often prominent in medieval spirituality. Implicit in this particular rejection of the traditional hierarchy of faculties in the personality, however, was a radical rejection of the traditional belief that hierarchy was the basis of all order. For Calvin, instead, the only foundation for order in human affairs was utility. Among its other consequences this position undermined the traditional one subordinating women to men. Calvin believed that, for practical reasons, it may be necessary for some to command and others to obey, but it could no longer be argued that women must naturally be subordinated to men. This helps to explain the rejection in Geneva of the double standard in sexual morality.

Conception of human personality

Second, Calvin's utilitarianism, as well as his understanding of the human personality as both less and more than intellectual, was also reflected in deep reservations about the capacity of human beings for anything but practical knowledge. The notion that they can know anything absolutely, as God knows, so to speak, seemed to him highly presumptuous. This conviction helps to explain his reliance on the Bible. Calvin believed that human beings have access to the saving truths of religion only insofar as God has revealed them in Scripture. But revealed truths were not given to satisfy human curiosity but were limited to meeting the most urgent and practical needs of human existence, above all for salvation. This emphasis on practicality reflects a basic conviction of Renaissance

Stress on practicality

humanism: the superiority of an active earthly life devoted to meeting practical needs to a life of contemplation. Calvin's conviction that every occupation in society is a "calling" on the part of God himself sanctified this conception. Calvin thus spelled out the theological implications of Renaissance humanism in various ways.

But Calvin was not purely a Renaissance humanist. The culture of the 16th century was peculiarly eclectic, and, like other thinkers of his time, Calvin had inherited a set of contrary tendencies, which he uneasily combined with his humanism. He was an unsystematic thinker not only because he was a humanist but also because 16th-century thinkers lacked the historical perspective that would have enabled them to sort out the diverse materials in their culture. Thus, even as he emphasized the heart, Calvin continued also to think of the human personality in traditional terms as a hierarchy of faculties ruled by reason. He sometimes attributed a large place to reason even in religion and emphasized the importance of rational control over the passions and the body. The persistence of these traditional attitudes in Calvin's thought, however, helps to explain its broad appeal; they were reassuring to conservatives.

Theology. Calvin has often been seen as little more than a systematizer of the more creative insights of Luther. He followed Luther on many points: on original sin, Scripture, the absolute dependence of human beings on divine grace, and justification by faith alone. But Calvin's differences with Luther are of major significance, even though some were largely matters of emphasis. Calvin was thus perhaps more impressed than Luther by God's transcendence and by his control over the world; Calvin emphasized God's power and glory, whereas Luther often thought of God as the babe in the manger, here among human beings. Contrary to a general impression, Calvin's understanding of predestination was also virtually identical with Luther's (and indeed is close to that of Thomas Aquinas); and, although Calvin may have stated it more emphatically, the issue itself is not of central importance to his theology. He considered it a great mystery, to be approached with fear and trembling and only in the context of faith. Seen in this way, predestination seemed to him a comforting doctrine; it meant that salvation would be taken care of by a loving and utterly reliable God.

Calvin and Luther

But in major respects Calvin departed from Luther. In some ways Calvin was more radical. Though he agreed with Luther on the real presence of Christ in the Eucharist, he understood this in a completely spiritual sense. But most of his differences suggest that he was closer to the old church than was Luther, as in his ecclesiology, which recognized the institutional church in this world, as Luther did not, as the true church. He was also more traditional in his clericalism; his belief in the authority of clergy over laity was hardly consistent with Luther's stress on the priesthood of all believers. He insisted, too, on the necessity of a holy life, at least as a sign of genuine election. Even more significant, especially for Calvinism as a historical force, was Calvin's attitude toward the world. Luther had regarded this world and its institutions as incorrigible and was prepared to leave them to the Devil, a far more important figure in his spiritual universe than in Calvin's. But for Calvin this world was created by God and still belonged to him. It was still potentially Christ's kingdom, and every Christian was obligated to struggle to make it so in reality by bringing it under God's law.

Attitude toward the world

Spirituality. Calvin's reservations about the capacities of the human mind and his insistence that Christians exert themselves to bring the world under the rule of Christ suggest that it is less instructive to approach his thought as a theology to be comprehended by the mind than as a set of principles for the Christian life—in short, as spirituality. His spirituality begins with the conviction that human beings do not so much "know" God as "experience" him indirectly, through his mighty acts and works in the world, as they experience but can hardly be said to know thunder, one of Calvin's favourite metaphors for religious experience. Such experience of God gives them confidence in his power and stimulates them to praise and worship him.

At the same time that Calvin stressed God's power, he also depicted God as a loving father. Indeed, although Calvinism is often considered one of the most patriarchal forms of Christianity, Calvin recognized that God is commonly experienced as a mother. He denounced those who represent God as dreadful; God for him is "mild, kind, gentle, and compassionate." Human beings can never praise him properly, Calvin declared, "until he wins us by the sweetness of his goodness." That God loves and cares for his human creatures was, for Calvin, what distinguished his doctrine of providence from that of the Stoics.

Calvin's understanding of Christianity is thus in many ways gentler than has been commonly supposed. This is also shown in his understanding of original sin. Although he insisted on the "total depravity" of human nature after the Fall, he did not mean by this that there is nothing good left in human beings but rather that there is no agency within the personality left untouched by the Fall on which to depend for salvation. The intention of the doctrine is practical: to reinforce dependence on Christ and the free grace of God. In fact, unlike some of his followers, Calvin believed in the survival after the Fall, however weak, of the original marks of God's image, in which human beings were created. "It is always necessary to come back to this," he declared, "that God never created a man on whom he did not imprint his image." At times, to be sure, Calvin's denunciations of sin give a very different impression. But it should be kept in mind that as a humanist and a rhetorician Calvin was less concerned to be theologically precise than to impress his audience with the need to repent of its sins.

Total depravity

The problem posed by sin was, for Calvin, not that it had destroyed the spiritual potentialities of human beings but rather that human beings had lost their ability to use their potentialities. Through the Fall they had been alienated from God, who is the source of all power, energy, warmth, and vitality. Sin, on the contrary, had exposed the human race to death, the negation of God's life-giving powers. Human beings thus experience the effects of sin as drowsiness when they should be alert, as apathy when they should feel concern, as sloth when they should be diligent, as coldness when they should be warm, as weakness when they need strength. Thus also, since the Devil, who seeks to drain human beings of their God-given spirituality, tries to lull them to sleep, God must employ various stratagems to awaken them. This helps to explain the troubles that afflict the elect: God threatens, chastises, and compels them to remember him by making their lives go badly.

The effect of sin also prevents human beings from reacting with appropriate wonder to the marvels of the world. The failure of spirituality is the primary obstacle to an affective knowledge that, unlike mere intellectual apprehension, can move the whole personality. Calvin attached particular importance to the way in which sin deadens the feelings, but spiritual knowledge renews the connection, broken by sin, between knowledge, feeling, and action. Thus God's spirit, in all its manifestations, is the power of life. Calvin's understanding of sin is closely related to his humanistic emphasis on activity.

As his emphasis on sanctification for the individual believer and on reconquering the world for Christ implies, Calvin's spirituality also included a strong sense of history, which he perceived as a process in which God's purposes are progressively realized. Therefore, the central elements of the Gospel—the Incarnation and Atonement, the grace available through them, the gift of faith by which human beings are enabled to accept this grace for themselves, and the sanctification that results—together describe objectively how human beings are enabled, step by step, to recover their original relationship with God and regain the energy coming from it. Calvin described this as a "quickening" that, in effect, brings the believer back from death to life and makes possible the most strenuous exertion in God's service.

Calvin exploited two traditional metaphors for the life of a Christian. Living in an unusually militant age, he drew on the familiar idea of the believer's life as a ceaseless, quasi-military struggle against the powers of evil both within the self and in the world. The Christian, in this conception, must struggle against his own wicked impulses, against the majority of the human race on behalf of the Gospel, and ultimately against the Devil. Paradoxically, however, Christian warfare consists less in inflicting wounds on others than in suffering the effects of sin patiently, that is, by bearing the cross. In Calvin's thought the metaphor for the Christian life as conflict thus takes on the added meaning of acquiescence in suffering. The disasters that afflict human existence, though punishments for the wicked, are an education for the believer; they strengthen faith, develop humility, purge wickedness, and compel him to keep alert and look to God for help.

The second traditional metaphor for the Christian life employed by Calvin, that of a journey or pilgrimage—*i.e.,* of a movement toward a goal—equally implied activity. "Our life is like a journey," Calvin asserted; yet "it is not God's will that we should march along casually as we please, but he sets the goal before us, and also directs us on the right way to it." This way is also a struggle because no one moves easily forward and most are so weak that, "wavering and limping and even creeping along the ground, they move at a feeble pace." Yet with God's help everyone can daily make some advance, however slight. Notable in this conception is a single-mindedness often associated with Calvinism: Christians must look straight ahead to the goal and be distracted by nothing, looking neither to the right nor left. Calvin allows them to love the good things in this life, but only within limits.

The progress of Christian life

Thus the Christian life is a strenuous progress in holiness, which, through the constant effort of the individual to make the whole world obedient to God, will also be reflected in the progressive sanctification of the world. These processes, however, will never be completed in this life. For Calvin even the most developed Christian in this world is like an adolescent, yearning to grow into, though still far from, the full stature of Christ. But, Calvin assured his followers, "each day in some degree our purity will increase and our corruption be cleansed as long as we live in the world," and "the more we increase in knowledge, the more should we increase in love." Meanwhile the faithful experience a vision, always more clear, of "God's face, peaceful and calm and gracious toward us." So the spiritual life, for Calvin as for many before him, culminates in the vision of God.

CALVINISM

Developments after Calvin. While Lutheranism was largely confined to parts of Germany and to Scandinavia, Calvinism spread into England, Scotland, the English-speaking colonies in North America, France, the Netherlands, much of Germany, and parts of central Europe. This expansion began during Calvin's lifetime and was encouraged by him. Religious refugees had poured into Geneva, especially from France during the 1550s as the French government became increasingly intolerant but also from England, Scotland, Italy, and other parts of Europe into which Calvinism had spread. Calvin welcomed them, trained many of them as ministers, sent them back to their countries of origin to spread the Gospel, and then supported them with letters of encouragement and advice. Geneva thus became the centre of an international movement and a model for churches elsewhere. John Knox, the Calvinist leader of Scotland, described Geneva as "the most perfect school of Christ that ever was on the earth since the days of the Apostles."

The appeal of Calvinism

Efforts to explain the appeal of Calvinism in social terms have had only limited success. In France it may have been primarily attractive to the nobility and the urban upper classes, in Germany it found adherents among both townsmen and princes, but in England and the Netherlands it made converts in every social group. It seems likely, therefore, that its appeal was based on its ability to explain disorders of the age afflicting all classes as well as on the remedies and comfort provided both by its activism and by its doctrine.

Having said this much, however, it is important to observe that the later history of Calvinism has often been obscured by a failure to distinguish between (1) Calvinism as the beliefs of Calvin himself, as discussed above, (2) the beliefs of his followers, who, though striving to be faithful to Calvin, modified his teachings to meet their own needs, and (3) more loosely, the beliefs of the Reformed tradition of Protestant Christianity, in which Calvinism proper was only one, if historically the most prominent, strand. The Reformed churches—in the 16th century referred to in the plural to indicate, along with what they had in common, their individual autonomy and variety—consisted originally of a group of non-Lutheran Protestant churches based in towns in Switzerland and southern Germany. These churches have always been jealous of their autonomy, and Geneva was not alone among them in having distinguished theological leadership. Huldrych Zwingli and Heinrich Bullinger in Zürich and Martin Bucer in Strasbourg also had a European influence, which combined with that of Calvin, especially in England, to shape what came to be called Calvinism.

The church in Geneva continued to venerate Calvin and aimed to be faithful to his teaching under his successors, first among them Theodore Beza, Calvin's chief lieutenant during the latter part of his life. But in the late 16th and early 17th centuries, in the atmosphere of what can be appropriately described as a Protestant "Counter-Reformation," Calvinism in Geneva underwent a change. Abandoning Calvin's more humanistic tendencies and drawing more on other aspects of his thought, Calvinism was increasingly intellectualized and came more and more to resemble the Scholasticism that Calvin had abhorred.

Under the influence of Aristotle the theology of Geneva became increasingly systematic. Faith, in this new atmosphere, was less a lively trust in God's promises than assent to a body of theological propositions. Especially the doctrine of predestination began to assume an importance such as had not been attributed to it before. Whereas Calvin had been led by personal faith to an awed belief in predestination, predestination, considered an "eternal decree" of God and a metaphysical necessity, now became the basis of faith.

Synod of Dort

Developments in Geneva illustrate what happened to Calvinism elsewhere. In 1619 they reached a climax at the Synod of Dort in the Netherlands, which spelled out various corollaries of predestination, as Calvin had never done, and made the doctrine central to Calvinism. Although the controversy that provoked this formulation was local, the synod was attended by representatives of Reformed churches elsewhere and assumed somewhat the same importance for them as the Council of Trent did for Roman Catholics.

In keeping with these developments Calvinist theologians, apparently finding Calvin's loose rhetorical style of expression unsatisfactory, began deliberately to write like Scholastic theologians, in Latin, and even appealed to medieval Scholastic authorities. The major Calvinist theological statement of the 17th century was the *Institutio Theologiae Elencticae* (1688) of Francois Turretin, chief pastor of Geneva. Although the title of his work recalled Calvin's masterpiece, the work itself bore little resemblance to the *Institutes;* it was not published in the vernacular, its dialectical structure followed the model of the great *Summae* of Thomas Aquinas, and it suggested Thomas' confidence in the value of human reason. The lasting significance of this shift is suggested by the fact that Turretin, in Latin, was the basic textbook in theology at the Princeton Theological Seminary in New Jersey, the most distinguished intellectual centre of American Calvinism, until the middle of the 19th century.

Historians of Calvinism have continued to debate whether these developments were essentially faithful to the beliefs of Calvin or deviations from them. In some sense they were both. Later Calvinism, though abandoning Calvin's more humanistic tendencies, found precedents for these changes in the contrary aspects of his thought. They were untrue to Calvin, however, in rejecting his typically Renaissance concern to balance contrary impulses. These changes, moreover, suggest the stage in the development of a movement that Max Weber called "routinization." It is the stage that comes after a movement's creative beginnings and, as a kind of reaction against the disorderly freedom of individual creativity, represents the quite different values of order and regularity. It is also relevant to explaining these changes in Calvinism that they occurred during a period of singular disorder, caused among other things by a century of religious warfare, which generally produced a longing for certainty, security, and peace.

Assessment. Calvin's influence has persisted not only in the Reformed churches of France, Germany, Scotland, the Netherlands, and Hungary but also in the Church of England, where Calvin was long at least as highly regarded as among those Puritans who separated from the Anglican establishment. The latter organized their own churches, Presbyterian or Congregational, which brought Calvinism to North America. Even today these churches, along with the originally German Evangelical and Reformed Church, recall Calvin as their founding father. Eventually Calvinist theology was also widely accepted by major groups of Baptists; and even Unitarianism, which broke away from the Calvinist churches of New England in the 18th century, reflected the more rational impulses in Calvin's theology. More recently Protestant interest in the social implications of the Gospel and Protestant neo-orthodoxy, as represented by Karl Barth, Emil Brunner, and Reinhold Niebuhr, reflects the continuing influence of John Calvin.

Calvin's larger influence over the development of modern Western civilization has been variously assessed. The controversial "Weber thesis" attributed the rise of modern capitalism largely to Puritanism, but neither Max Weber, in his famous essay of 1904, "Die protestantische Ethik und der Geist des Kapitalismus" (*The Protestant Ethic and the Spirit of Capitalism*), nor the great economic historian Richard Henry Tawney, in *Religion and the Rise of Capitalism* (1926), implicated Calvin himself in this development. Much the same thing can be said about efforts to link Calvinism to the rise of modern science; although Puritans were prominent in the scientific movement of 17th-century England, Calvin himself was indifferent to the science of his own day. A somewhat better case can be made for Calvin's influence on political theory. His own political instincts were highly conservative, and he preached the submission of private persons to all legitimate authority. But, like Italian humanists, he personally preferred a republic to a monarchy. In confronting the problem posed by rulers who actively opposed the spread of the Gospel, he advanced a theory of resistance, kept alive by his followers, according to which lesser magistrates might legitimately rebel against kings. Unlike most of his contemporaries, furthermore, Calvin included among the proper responsibilities of states not only the maintenance of public order but also a positive concern for the general welfare of society.

Calvinism has a place, therefore, in the development of liberal political thought. Calvin's major and most durable influence, nevertheless, has been religious. From his time to the present Calvinism has meant a peculiar seriousness about Christianity and its ethical implications.

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(W.J.Bo.)

Canada

The second largest country in the world in area (after the Soviet Union) but one of the most sparsely populated, Canada occupies roughly two-fifths of the North American continent. Its total area, including the Canadian share of the Great Lakes, is 3,849,672 square miles (9,970,610 square kilometres), of which 291,576 square miles (755,180 square kilometres) are inland water; also included are its adjacent islands, except Greenland, a self-governing part of the Danish kingdom, and Saint-Pierre and Miquelon, parts of the French Republic. Canada is bounded on the north by the Arctic Ocean, on the east by the Atlantic Ocean, on the south by 12 states of the United States, and on the west by the Pacific Ocean and the U.S. state of Alaska. The national capital for the 10 provinces and two territories of Canada is Ottawa.

Canada shares a 5,527-mile (8,895-kilometre) border with the United States. In longitude Canada extends from approximately 52° W to 141° W, a distance that spans six time zones. In latitude it extends from approximately 42° N to 83° N. With its vast Arctic and subarctic territories, Canada is often considered a country only of the far north. It can be noted, however, that the peninsula of southern Ontario juts deeply into the heartland of the United States, its southernmost point, Pelee Island, being south of 11 states of the United States. Canada has a strategic global location. It lies on great circle routes (the shortest line joining any two places on the globe) between the United States and Europe and, to a lesser degree, Asia. For this reason many international commercial air flights track across Canada.

Most of Canada's farmland and densely populated areas are located along a narrow strip that rims the U.S. border. In the highly urbanized areas of southern Ontario and Quebec, the population density compares with the more densely populated areas of the United States and Europe. More than 75 percent of Canada's population lives in urban centres.

Canada shares the largest sector of the North American continent with the United States, and there is considerable U.S. influence in Canada. It is apparent, in fact, to travelers crossing the border that the two countries are similar in many ways. The landscape does not change, and the buildings, farms, and dress are all comparable. Neither is there much variation in foods and the kinds of products available. Some differences do stand out, however. The type of government, the political organization, and the judiciary all display some marked contrasts, resembling British institutions rather than American. Education in Canada reflects both British and U.S. influence, although Quebec in particular shows some French adaptations. The French presence in Canada is a unique characteristic. More than three-fourths of Quebec's people have French as their primary language. The French character in Quebec is reflected in the architecture as well as the language. Throughout most of the rest of the country, the French influence is apparent in the dual use of French for place-names, product labeling, and street signs, in schools that teach in French, and in French-language radio and television programs. In addition, increased immigration from other European countries, Southeast Asia, and Latin America in the latter part of the 20th century has made Canada even more broadly multicultural. Added to the spectrum are the cultures of the native Indians and the Inuit, the former being far greater in number. Inuit, rather than Eskimo, is the name preferred by the latter and is the term of common use in Canada.

The name Canada, which is the official name of the country, is spelled the same in the country's two official languages, English and French. The word Canada is a derivation of the Huron-Iroquois *kanata,* meaning a village or settlement. The explorer Jacques Cartier used the name Canada to refer to the area around the settlement which is now Quebec city. Later, Canada was used as a synonym for New France, which included all the French possessions along the St. Lawrence River and the Great Lakes. After the British conquest of New France, the name Quebec was used for a while instead of Canada. The name Canada was restored after 1791, when Britain divided Quebec into the provinces of Upper and Lower Canada. In 1867 the British North America Act created a confederation of colonies called the Dominion of Canada. The word Dominion fell into disuse, but the name Canada has stood the test of time.

Canada has long been an influential member of the Commonwealth and has played a leading role in the organization of French-speaking countries known as Francophonie. Canada is a founding member of the United Nations (UN) and has been active in a number of major UN agencies and other worldwide operations. In 1989 it joined the Organization of American States (OAS).

The combination of physical geography and discontinuous settlement has led to a strong sense of regionalism in Canada, and popular regional terms often overlap. The Atlantic Provinces refers to all of the Appalachian region except the Quebec portion. If Newfoundland is excluded, the three remaining east-coast provinces are called the Maritimes. Quebec and Ontario are usually referred to separately but sometimes together as Central Canada; the West usually means all of the four provinces west of Ontario, but British Columbia may be referred to alone and the other three collectively as the Prairies, or Prairie Provinces; the Yukon and Northwest Territories are referred to as the North.

The article is divided into the following sections:

PHYSICAL AND HUMAN GEOGRAPHY

The land

RELIEF

In general terms, the landform structure of Canada can be considered as a vast basin more than 3,220 miles in diameter. The Cordillera in the west, the Appalachians in the southeast, the mountains of northern Labrador and of Baffin Island in the northeast, and the Innuitian Mountains in the north form its high rim, while Hudson Bay, set close to the centre of the enormous platform of the Canadian Shield, occupies the basin bottom. The western rim of the basin is higher and more massive than its eastern counterpart, and pieces of the rim, notably in the far northwest and in the south, are missing. The main lines of Canadian landforms continue well into the United States, intimately linking the geography north and south of the border. To create a nation from sea to sea, Canada had to forge transportation and communication links in an east–west direction, against the physiographic grain of the continent. The Canadian North remains one of the least settled and least exploited parts of the world. Canada can be divided into several physical regions, including the Canadian Shield, the interior plains, the Great Lakes–St. Lawrence lowlands, the Appalachian region, the Western Cordillera, and the Arctic Archipelago.

The Canadian Shield. By far the largest of Canada's physiographic regions, the Canadian Shield occupies about half of the total area of the country and is centred about Hudson Bay. It consists of ancient rocks folded by mountain-building movements and cut down by erosion until the area was reduced almost to a plain. It was warped and folded in places so that parts of it now stand much higher than others, especially around its outer edges. In the north the rim is about 7,000 feet (2,000 metres) above sea level, and fjords with walls from 2,000 to 3,000 feet high extend many miles into the mountain masses. South of Hudson Strait lie the Labrador Highlands, including the Torngat, Kaumajet, and Kiglapait mountains. Along the north shore of the St. Lawrence River in Quebec, the shield rim is a 2,000-foot escarpment—the Laurentide Scarp. The rim is almost imperceptible in southern Ontario, but in northern Ontario it rises again to almost 1,500 feet above the northern shore of Lake Superior. From Manitoba northwestward the shield edge is marked by a number of lakes.

Janet Foster/Masterfile

Nachvak Fjord in the Torngat Mountains, Labrador, the highest range in the Canadian Shield.

Formation of the landscape

Most of the shield itself is less than 2,000 feet above sea level. Its lack of hills of any size produces a generally monotonous landscape, but geologically recent glaciations have had a striking effect on the surface. By stripping off the top, weathered material, they roughened the surface into a type of rock-knob, or grained, landscape, with the hollows between the knobs or the troughs between the ridges occupied by enormous numbers of lakes. In other areas the glaciers deposited till or moraine on the surface and in still others left gigantic fields of erratics and boulders. Eskers—long, narrow ridges of deposits—stretch across the shield sometimes for more than 100 miles, marking the course of old, subglacial rivers. In still other places, deposits laid down by glacial lakes that have since drained away have given rise to extensive clay belts.

The interior plains. Surrounding the Canadian Shield are a number of extensive lowlands underlain by sedimentary rocks: the Arctic lowlands to the north, the Great Lakes–St. Lawrence lowlands to the south and southeast, and the interior, or western, plains to the west. The most extensive of these are the interior plains, which extend from the Arctic Ocean in the north to the United States border in the south and from the edge of the Canadian Shield in the east to the Rocky Mountain foothills in the west. Along the shield–interior plains boundary are a number of major lakes, three of which are larger than Lake Ontario: Great Bear, Great Slave, and Winnipeg.

In the southeast is the Manitoba lowland, with an elevation of less than 1,000 feet above sea level. It is underlain by lacustrine sediments of the glacial Lake Agassiz and is the flattest land in the interior plains. It includes Lake Manitoba and Lake Winnipegosis in addition to Lake Winnipeg. The fertile southern portion, known as the Red River valley, is covered with black clay and silt soils.

To the west of the Manitoba lowland, the land rises in two steps: the Saskatchewan plain (from 1,500 to 2,100 feet above sea level) and the Alberta plain (more than 2,500 feet above sea level). These plains are rolling landscapes of glacial deposits laid over an almost horizontal bedrock. In some areas the undulating plains are interspersed with ranges of low hills (glacial moraines) studded with kettle lakes and flat-bottomed, steep-banked valleys cut by glacial meltwater, now occupied by rivers such as the Assiniboine, Saskatchewan, and Qu'Appelle. Ponds called sloughs dot the landscape of both these plains. In Saskatchewan, the Cypress Hills rise 4,816 feet above the surrounding plain including that portion lying in Alberta, to the highest point in mainland Canada between the Rocky Mountains and Labrador.

The Mackenzie Lowlands, extending from the Alberta

The Mackenzie Lowlands

plain north to the Arctic Ocean, is a flat area covered with muskeg (bogs) and swamps. It is drained by the Mackenzie River.

The Great Lakes–St. Lawrence lowlands. The Great Lakes–St. Lawrence region comprises the peninsula of southern Ontario bounded by the Canadian Shield and Lakes Huron, Erie, and Ontario. It extends along the St. Lawrence River to the Atlantic Ocean. This is a much smaller region, but it is important for its high agricultural productivity, its intensive industrialization, and its high degree of urbanization.

The rolling landscape of the Great Lakes–St. Lawrence lowlands is composed primarily of glacial landforms: glacial lake bottoms and shorelines, till plains, moraines, drumlins, eskers, and giant spillways carved by glacial streams. In southwestern Ontario the Niagara Escarpment is the only significant exposed bedrock structure. This steep cuestaform ridge runs from Niagara Falls to the Bruce Peninsula west of Georgian Bay and on into Manitoulin Island. In southeastern Ontario the lowland is interrupted by a band of the Canadian Shield, the Frontenac Axis, which extends across the St. Lawrence River to form the Thousand Islands.

Northeast of the Frontenac Axis, the lowlands embrace the Ottawa valley and the St. Lawrence valley to a point some 70 miles downstream from Quebec city. During the last glacial period, this area was inundated by ocean water, known as the Champlain Sea, resulting in a very flat plain. The level plain is broken by the seven Monteregian Hills near Montreal. The westernmost of these is Mont-Royal in Montreal, about 820 feet high.

The Appalachian region. The Appalachian region extends from the eastern townships of Quebec (south of the St. Lawrence valley) northeastward to the Gaspé Peninsula and the Maritime Provinces, and on to the island of Newfoundland. The region consists of very old folded rock formations that have been eroded into low, rounded mountains dissected by valleys and interrupted by lowland areas developed on weaker rock formations. Three broad groups of highlands can be recognized. The highest mountains, with elevations exceeding 4,000 feet above sea level, are found in southern Quebec: Sutton, Notre-Dame, and Shickshock mountains. The highlands in New Brunswick and Nova Scotia are lower, and the hills have been dissected out of a plateau upland. The major portion of Newfoundland is also a dissected plateau, but along the east coast, the Long Range Mountains rise to more than 2,000 feet in elevation. The relatively small areas of lowland of the region extend along the seacoast and the major rivers.

Highland division of the Appalachian region

The Western Cordillera. The Cordilleran region comprises a series of mountain belts some 500 miles wide that flanks the Pacific coast of Canada. The great heights and angularity of the peaks (many of which rise to more than 10,000 feet) indicate that these are much younger mountains than the Appalachians. Signs of alpine glaciation are widely evident. In many places valley glaciers remain active, and snowcapped peaks are frequently hidden in the clouds. Some of the mountain slopes are so precipitous that they are bare of trees. Viewed from an airplane, the whole landscape seems to be an irregular sea of mountain ranges, trending in a north–south direction.

The Cordilleran region is composed of numerous mountain groups, the major ones of which are discussed here. The Rocky Mountains make up the eastern edge of the Cordillera from the Yukon border south to the 49th parallel, where they continue into the United States. They form the continental divide between eastward- and westward-flowing rivers and contain some of the most rugged and scenic landscape in North America. The highway between Banff and Jasper is particularly noted for its spectacular mountain scenery. The Rockies include 30 peaks more than 10,000 feet high, among them Mount Robson, which rises to 12,972 feet. Five of Canada's national parks are located within the Rockies, including Banff, which was established in 1885. Three major passes cut through the Rockies: the Yellowhead Pass, which is used by the Canadian National Railways, and the Kicking Horse Pass and Crowsnest Pass, which are used by the Canadian Pacific Railway. The Trans-Canada Highway is routed through the Kicking Horse Pass.

West of the Rockies

To the west the Rocky Mountains are bordered by a major valley about 15 miles wide and several thousand feet deep, known as the Rocky Mountain Trench. West of the Rocky Mountain Trench the Columbia Mountain system rises to peaks of more than 10,000 feet. The Columbia Mountains include, from east to west, the Purcell, Selkirk, and Monashee groups. Northwest of these are the Cariboo Mountains, world-famous for their helicopter alpine skiing. Between the Columbia Mountains and the Coast Mountains farther west is a broad region of interior mountains and plateaus. Although some of the surface of this region is fairly level, most of it has been dissected into mountains and hills.

The Coast Mountains are another group of high mountains, with several peaks rising to over 15,000 feet, including Canada's highest peak, Mount Logan, rising to 19,524 feet (5,951 metres) in the St. Elias Mountains. All along the coast there are spectacular fjords with precipitous cliffs that often rise 7,000 feet from the water. Off the coast is a chain of mountains that appear as a series of islands, the largest of which are Vancouver Island and the Queen Charlotte Islands. In the far north the main mountain groups are the Richardson, Mackenzie, Selwyn, and Pelly mountains. The rugged Cassiar Mountains stand just south of the Yukon border.

The Arctic Archipelago. The Arctic Archipelago is composed of thousands of islands north of the Canadian mainland. The southeastern islands are an extension of the Canadian Shield. The balance consists of two distinctive landform regions: the Arctic lowlands to the south and the Innuitian Mountains to the north. The Innuitians are geologically young mountains similar to the Western Cordillera, with peaks reaching 10,000 feet. Much of the Innuitian region is permanently covered with snow and ice through which mountain peaks occasionally protrude.

DRAINAGE

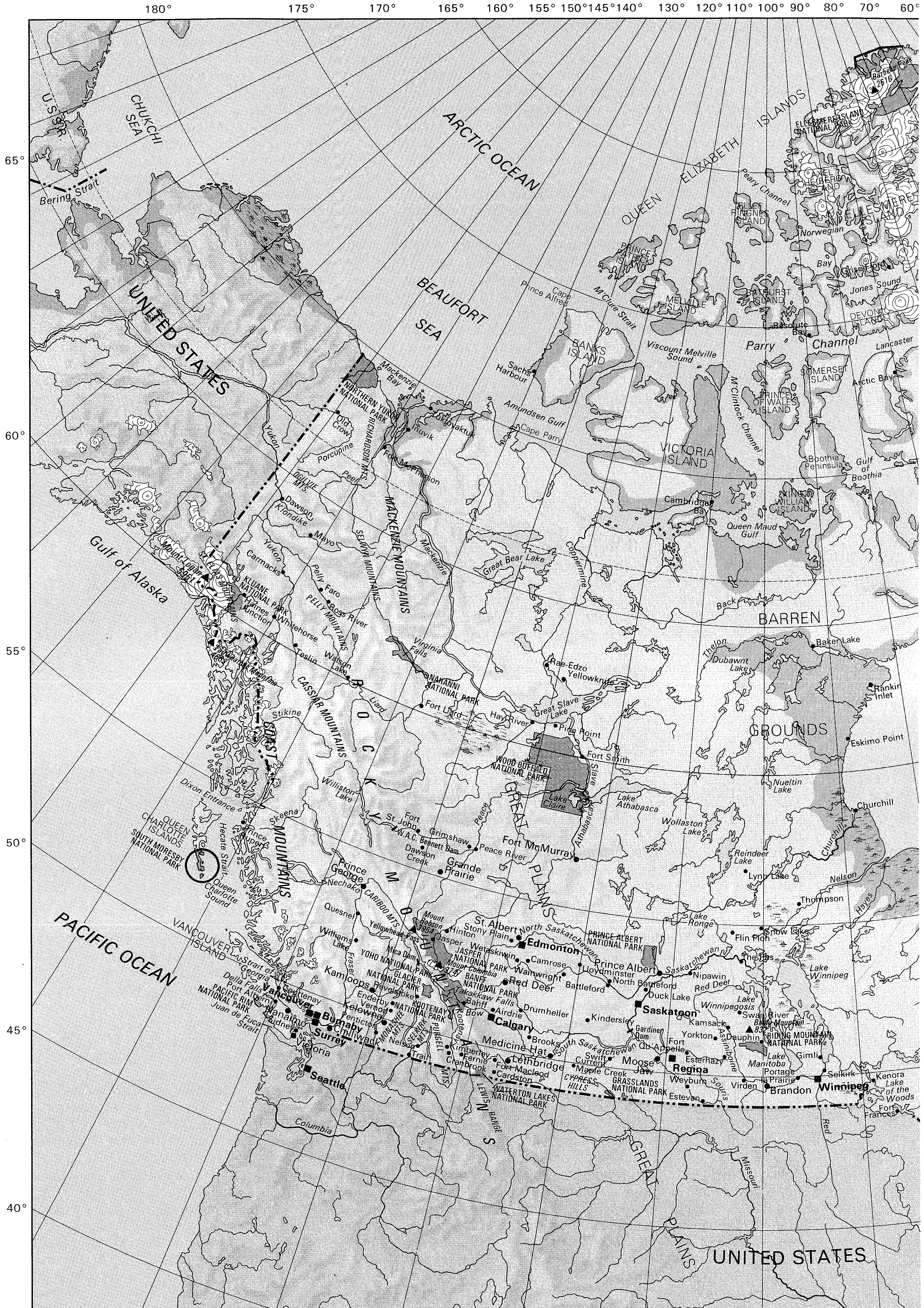
With just 1 percent of the world's population, Canada has some 9 percent of the world's supply of fresh water. Much of this water is stored in lakes and wetlands that cover about 20 percent of Canada's total area. About three-quarters of Canada's land area is drained by rivers flowing into the Arctic Ocean and the Hudson and James bays. The Arctic drainage basin is dominated by the Mackenzie River, Canada's longest river, which flows 2,635 miles from its source to its mouth. With its many tributaries, it drains 690,000 square miles. The St. Lawrence is the largest river flowing into the Atlantic Ocean. Its drainage basin includes the Great Lakes, forming an inland navigable waterway extending 1,900 miles into the heart of the continent. The longest Pacific-draining river that is wholly within Canada is the Fraser; the Yukon and Columbia rivers also flow to the Pacific.

The St. Lawrence River

The usefulness of Canadian rivers is limited by two factors: many flow through the northern part of the country, which is sparsely populated, and most of them are frozen over in winter. In the densely settled regions, pollution has further reduced the usefulness of the water. Almost all Canadian rivers are characterized by rapids and falls, many of which have been developed for hydroelectricity.

CLIMATE

Because of its great latitudinal extent, Canada has a wide variety of climates. The Pacific coast, like that of the U.S. Northwest coast, has a climate similar to that of England, with heavy precipitation and moderate winter and summer temperatures. The northern two-thirds of the country has a climate similar to that of northern Scandinavia, with very cold winters and short, cool summers. The central southern area of the interior plains has a typical continental climate—very cold winters, hot summers, and relatively sparse precipitation. Southern Ontario and Quebec have a climate with hot, humid summers and cold, snowy winters, similar to that of some portions of the American Midwest. All of Canada, except the west coast, has a winter season with average temperatures below freezing and with continuous snow cover.


180°
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95°
U.S.S.R.
CHUKCHI SEA
Bering Strait
ARCTIC OCEAN
QUEEN ELIZABETH ISLANDS
ELLESMERE ISLAND NATIONAL PARK
Barbeau Peak 2616
AXEL HEIBERG ISLAND
ELLESMERE ISLAND
Norwegian Bay
Grise Fiord
Jones Sound
DEVON ISLAND
Peary Channel
ELLEF RINGNES ISLAND
PRINCE PATRICK ISLAND
MELVILLE ISLAND
BATHURST ISLAND
Resolute Bay
Parry Channel
Lancaster
BEAUFORT SEA
Prince Alfred
Cape
M'Clure Strait
Viscount Melville Sound
BANKS ISLAND
Sachs Harbour
SOMERSET ISLAND
Arctic Bay
PRINCE OF WALES ISLAND
M'Clintock Channel
Amundsen Gulf
Cape Parry
VICTORIA ISLAND
Boothia Peninsula
Gulf of Boothia
KING WILLIAM ISLAND
Cambridge Bay
Queen Maud Gulf
UNITED STATES
Mackenzie Bay
NORTHERN YUKON NATIONAL PARK
Old Crow
Porcupine
RICHARDSON MTS.
Inuvik
Tuktoyaktuk
Fort McPherson
Yukon
Peel
OGILVIE MTS.
Dawson
Klondike
Mayo
SELWYN MOUNTAINS
MACKENZIE MOUNTAINS
Mackenzie
Great Bear Lake
Coppermine
Gulf of Alaska
Mount Logan 5951
ST. ELIAS MOUNTAINS
KLUANE NATIONAL PARK
Haines Junction
Carmacks
Yukon
Pelly
Faro
PELLY MOUNTAINS
Ross River
Whitehorse
Teslin
Watson Lake
Virginia Falls
NAHANNI NATIONAL PARK
Back
BARREN
Thelon
Dubawnt Lake
Baker Lake
Rankin Inlet
Rae-Edzo
Yellowknife
CASSIAR MOUNTAINS
Liard
Fort Liard
Hay River
Great Slave Lake
Pine Point
GROUNDS
Eskimo Point
Stikine
COAST
Fort Smith
WOOD BUFFALO NATIONAL PARK
Slave
Lake Claire
R O C K Y
Williston Lake
GREAT PLAINS
Peace
Athabasca
Lake Athabasca
Nueltin Lake
Wollaston Lake
Churchill
Dixon Entrance
QUEEN CHARLOTTE ISLANDS
SOUTH MORESBY NATIONAL PARK
Hecate Strait
Prince Rupert
Skeena
MOUNTAINS
Fort St. John
W.A.C. Bennett Dam
Dawson Creek
Grimshaw
Peace River
Fort McMurray
Grande Prairie
Reindeer Lake
Lynn Lake
Churchill
Nelson
Thompson
Hayes
Queen Charlotte Sound
PACIFIC OCEAN
Prince George
Nechako
CARIBOO MTS.
Quesnel
M O U N T A I N S
Mount Robson 3954
Yellowhead Pass
Hinton
Jasper
Williams Lake
St. Albert
Stony Plain
Edmonton
North Saskatchewan
PRINCE ALBERT NATIONAL PARK
Lac la Ronge
Flin Flon
Snow Lake
The Pas
VANCOUVER ISLAND
Strait of Georgia
Mica Dam
JASPER NATIONAL PARK
Mount Columbia 3747
Wetaskiwin
Camrose
Wainwright
Lloydminster
Prince Albert
Saskatchewan
Nipawin
Lake Winnipeg
YOHO NATIONAL PARK
GLACIER NATIONAL PARK
Kamloops
Revelstoke
BANFF NATIONAL PARK
Red Deer
Battleford
North Battleford
Red Deer
Della Falls
Fort Alberni
PACIFIC RIM NATIONAL PARK
Juan de Fuca Strait
Courtenay
Vancouver
Nanaimo
Sidney
Victoria
Burnaby
Surrey
Chilliwack
Fraser
Enderby
Vernon
Kelowna
Penticton
MONASHEE MTS.
KOOTENAY NATIONAL PARK
SELKIRK MTS.
PURCELL MTS.
Nelson
Trail
Takakkaw Falls
Banff
Bow
Airdrie
Calgary
Drumheller
Kindersley
Duck Lake
Saskatoon
Lake Winnipegosis
Swan River
Baldy Mountain 832
Kamsack
Yorkton
Dauphin
RIDING MOUNTAIN NATIONAL PARK
Gardiner Dam
Assiniboine
Kootenay
Kimberley
Fernie
Cranbrook
Brooks
Medicine Hat
South Saskatchewan
Swift Current
Fort Qu'Appelle
Esterhazy
Lake Manitoba
Gimli
Lethbridge
Fort Macleod
Cardston
Maple Creek
CYPRESS HILLS
Moose Jaw
Regina
GRASSLANDS NATIONAL PARK
Weyburn
Souris
Portage la Prairie
Selkirk
Kenora
Lake of the Woods
Virden
Brandon
Winnipeg
Estevan
Fort Frances
Red
WATERTON LAKES NATIONAL PARK
LEWIS RANGE
Seattle
Columbia
GREAT PLAINS
Missouri
UNITED STATES


Scale 1:19,410,000
1 inch equals approx. 306 miles
0 100 200 300 400 mi
0 150 300 450 600 km
Cities over 1,000,000
Cities 100,000 to 1,000,000
Cities 25,000 to 100,000
Cities under 25,000
National capitals
Political subdivision capitals
Political subdivision names
International boundaries
Political subdivision boundaries
Canals
Dams
Bridges
Waterfalls
Glaciers
Swamps and marshes
National parks
Spot elevations in metres
(1 m = 3.28 ft)
Lambert Conformal Conic Projection
55° 50° 45° 40° 35° 30° 25° 20° 15°
90° 85° 80° 75° 70° 65°
80° 75° 70° 60° 55° 50° 45° 40° 35°
GREENLAND (DEN.)
ICELAND
Reykjavík
Arctic Circle
Nuuk
Kane Basin
Smith Sound
Baffin Bay
Sound
BYLOT ISLAND
Barnes Ice Cap
BAFFIN ISLAND
Penny Ice Cap
AUYUITTUQ NATIONAL PARK
Pangnirtung
PRINCE CHARLES ISLAND
Melville Peninsula
Nettilling Lake
Cumberland Sound
Foxe Basin
Amadjuak Lake
Iqaluit
Frobisher Bay
Davis Strait
SOUTHAMPTON ISLAND
Hudson Strait
Evans Strait
COATS ISLAND
Nouveau-Québec Crater 657
Cape Chidley
Mount Caubvick
TORNGAT MTS.
LABRADOR SEA
UNGAVA PENINSULA
Ungava Bay
LABRADOR
Kuujjuaq
Caniapiscau
Hudson Bay
Hamilton Inlet
Lake Melville
North West River
Happy Valley-Goose Bay
Smallwood Reservoir
Churchill Falls
Churchill
Schefferville
Labrador City
Wabush
Caniapiscau Reservoir
BELCHER ISLANDS
Cape Henrietta Maria
La Grande 4 Reservoir
La Grande 2 Reservoir
James Bay
Opinaca Reservoir
AKIMISKI ISLAND
Severn
Winisk
Ekwan
Attawapiskat
Albany
Moose
Nottaway
Rupert
Lake Mistassini
Chibougamau
Gagnon
Manicouagan Reservoir
Daniel Johnson Dam
Outardes
Sept-Îles
Port-Cartier
Strait of Belle Isle
LONG RANGE MTS.
GROS MORNE NATIONAL PARK
Fogo
Lewisporte
Gander
TERRA NOVA NATIONAL PARK
Bonavista
Trinity Bay
Springdale
Grand Falls
Buchans
Corner Brook
NEWFOUNDLAND
Carbonear
St. John's
Mount Pearl
Placentia
Avalon Peninsula
Trepassey
Grand Bank
Marystown
Channel-Port aux Basques
SAINT-PIERRE AND MIQUELON (FR.)
ANTICOSTI ISLAND
Gulf of St. Lawrence
Gaspé
Gaspé Peninsula
Murdochville
Percé
Matane
NOTRE DAME MTS.
Rimouski
Baie-Comeau
St. Lawrence
MADELEINE ISLANDS
Cabot Strait
CAPE BRETON HIGHLANDS NATIONAL PARK
Glace Bay
Sydney
Port Hawkesbury
Strait of Canso
SABLE ISLAND
Campbellton
Bathurst
Mount Carleton 820
Chatham
Summerside
Charlottetown
Pictou
New Glasgow
Edmundston
Grand Falls
Fredericton
Saint John
Moncton
Truro
Windsor
Dartmouth
Halifax
Bridgewater
Liverpool
Shelburne
Digby
Yarmouth
Bay of Fundy
FUNDY NATIONAL PARK
Lake Saint-Jean
Alma
Chicoutimi
Jonquière
Roberval
La Baie
Rivière-du-Loup
Charlesbourg
Beauport
Québec
Sainte-Foy
Lévis
Montmagny
La Tuque
Shawinigan
Trois-Rivières
Cap-de-la-Madeleine
Joliette
Sherbrooke
Mont-Laurier
Montréal
Longueuil
Granby
Châteauguay
Hull
Ottawa
Nepean
Cornwall
Pembroke
Outaouais
Senneterre
Amos
Rouyn-Noranda
Val-d'Or
Abitibi
Kapuskasing
Iroquois Falls
Timmins
Kirkland Lake
New Liskeard
Lake Nipigon
PUKASKWA NATIONAL PARK
Thunder Bay
Lake Superior
Sault Ste. Marie
Sault Ste. Marie Bridge
Sudbury
Nickel Centre
Sturgeon Falls
North Bay
Elliot Lake
MANITOULIN ISLAND
Georgian Bay
Parry Sound
Orillia
Peterborough
Perth
Kingston
Belleville
Lake Ontario
Owen Sound
Barrie
Lake Huron
Toronto
Oshawa
Scarborough
Mississauga
Burlington
Kitchener
Hamilton
Niagara Falls
Welland
London
St. Thomas
Sarnia
Chatham
Lake St. Clair
Detroit
Windsor
Lake Erie
Cleveland
Lake Michigan
Mississippi
Milwaukee
Chicago
Lake Champlain
Boston
Hudson
APPALACHIAN MOUNTAINS
New York City
Philadelphia
Baltimore
Washington D.C.
ATLANTIC OCEAN
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MAP INDEX

Political subdivisions

Alberta	55 00 N	115 00 W
British Columbia	55 00 N	125 00 W
Manitoba	55 00 N	97 00 W
New Brunswick	47 00 N	66 00 W
Newfoundland	53 00 N	60 00 W
Northwest Territories	65 00 N	95 00 W
Nova Scotia	45 00 N	64 00 W
Ontario	50 00 N	85 00 W
Prince Edward Island	46 15 N	63 00 W
Quebec	53 00 N	70 00 W
Saskatchewan	55 00 N	105 00 W
Yukon Territory	65 00 N	135 00 W

Cities and towns

Airdrie	51 18 N	114 02 W
Alma	48 32 N	71 40 W
Amos	48 35 N	78 07 W
Arctic Bay	73 02 N	85 11 W
Baie-Comeau	49 13 N	68 09 W
Baker Lake	64 15 N	96 00 W
Banff	51 10 N	115 34 W
Barrie	44 24 N	79 40 W
Bathurst	47 36 N	65 39 W
Battleford	52 44 N	108 19 W
Beauport	46 52 N	71 11 W
Belleville	44 10 N	77 23 W
Bonavista	48 39 N	53 07 W
Borden	46 15 N	63 42 W
Brandon	49 50 N	99 57 W
Bridgewater	44 23 N	64 31 W
Brooks	50 35 N	111 53 W
Buchans	48 49 N	56 52 W
Burlington	43 19 N	79 47 W
Burnaby	49 16 N	122 57 W
Calgary	51 03 N	114 05 W
Cambridge Bay	69 03 N	105 05 W
Campbellton	48 00 N	66 40 W
Camrose	53 01 N	112 50 W
Cap-de-la-Madeleine	46 22 N	72 31 W
Carbonear	47 44 N	53 13 W
Cardston	49 12 N	113 18 W
Carmacks	62 05 N	136 17 W
Channel-Port aux Basques	47 34 N	59 09 W
Charlesbourg	46 51 N	71 16 W
Charlottetown	46 14 N	63 08 W
Châteauguay	45 23 N	73 45 W
Chatham	42 24 N	82 11 W
Chatham	47 02 N	65 28 W
Chibougamau	49 55 N	74 22 W
Chicoutimi	48 26 N	71 04 W
Chilliwack	49 10 N	121 57 W
Churchill	58 46 N	94 10 W
Churchill Falls	53 33 N	64 01 W
Corner Brook	48 57 N	57 57 W
Cornwall	45 02 N	74 44 W
Courtenay	49 41 N	124 59 W
Cranbrook	49 30 N	115 46 W
Dartmouth	44 40 N	63 34 W
Dauphin	51 09 N	100 03 W
Dawson	64 04 N	139 26 W
Dawson Creek	55 46 N	120 14 W
Digby	44 37 N	65 36 W
Dominion	46 13 N	60 01 W
Drumheller	51 28 N	112 42 W
Duck Lake	52 49 N	106 14 W
Edmonton	53 33 N	113 28 W
Edmundston	47 22 N	68 20 W
Elliot Lake	46 23 N	82 42 W
Enderby	50 33 N	119 09 W
Eskimo Point	61 07 N	94 03 W
Esterhazy	50 39 N	102 05 W
Estevan	49 08 N	102 59 W
Faro	62 14 N	133 20 W
Fernie	49 30 N	115 04 W
Flin Flon	54 46 N	101 53 W
Fogo	49 43 N	54 17 W
Fort-Chimo, see Kuujjuaq		
Fort Frances	48 36 N	93 24 W
Fort Liard	60 15 N	123 28 W
Fort Macleod	49 43 N	113 25 W
Fort McMurray	56 44 N	111 23 W
Fort McPherson	67 27 N	134 53 W
Fort Qu'Appelle	50 46 N	103 48 W
Fort St. John	56 15 N	120 51 W
Fort Smith	60 00 N	111 53 W
Fredericton	45 58 N	66 39 W
Frobisher Bay, see Iqaluit		
Gagnon	51 53 N	68 10 W
Gander	48 57 N	54 37 W
Gaspé	48 50 N	64 29 W
Gimli	50 38 N	96 59 W
Glace Bay	46 12 N	59 57 W
Granby	45 24 N	72 43 W
Grand Bank	47 06 N	55 46 W
Grand Falls	47 03 N	67 44 W
Grand Falls	48 56 N	55 40 W
Grande Prairie	55 10 N	118 48 W
Grimshaw	56 11 N	117 36 W
Grise Fiord	76 25 N	82 55 W
Haines Junction	60 45 N	137 30 W
Halifax	44 39 N	63 36 W
Hamilton	43 15 N	79 51 W
Happy Valley-Goose Bay	53 19 N	60 20 W
Harbour Grace	47 42 N	53 13 W
Hay River	60 49 N	115 47 W
Hinton	53 25 N	117 34 W
Hull	45 26 N	75 44 W
Inuvik	68 21 N	133 43 W
Iqaluit (Frobisher Bay)	63 45 N	68 31 W
Iroquois Falls	48 46 N	80 41 W
Jasper	52 53 N	118 05 W
Joliette	46 01 N	73 27 W
Jonquière	48 25 N	71 13 W
Kamloops	50 40 N	120 19 W
Kamsack	51 34 N	101 54 W
Kapuskasing	49 25 N	82 26 W
Kelowna	49 53 N	119 29 W
Kenora	49 47 N	94 29 W
Kimberley	49 41 N	115 59 W
Kindersley	51 28 N	109 10 W
Kingston	44 14 N	76 30 W
Kirkland Lake	48 09 N	80 02 W
Kitchener	43 27 N	80 29 W
Kuujjuaq (Fort-Chimo)	58 06 N	68 25 W
La Baie	48 20 N	70 52 W
La Tuque	47 26 N	72 47 W
Labrador City	52 57 N	66 55 W
Lethbridge	49 42 N	112 49 W
Lévis	46 48 N	71 11 W
Lewisporte	49 14 N	55 03 W
Liverpool	44 02 N	64 43 W
Lloydminster	53 17 N	110 00 W
London	42 59 N	81 14 W
Longueuil	45 32 N	73 30 W
Lynn Lake	56 51 N	101 03 W
Maple Creek	49 55 N	109 29 W
Marystown	47 10 N	55 09 W
Mayo	63 36 N	135 54 W
Medicine Hat	50 03 N	110 40 W
Mississauga	43 35 N	79 39 W
Moncton	46 07 N	64 48 W
Mont-Laurier	46 33 N	75 30 W
Montmagny	46 59 N	70 33 W
Montreal	45 30 N	73 36 W
Moose Jaw	50 24 N	105 32 W
Mount Pearl	47 31 N	52 47 W
Nanaimo	49 10 N	123 56 W
Nelson	49 30 N	117 17 W
Nepean	45 16 N	75 46 W
New Glasgow	45 35 N	62 39 W
New Liskeard	47 30 N	79 40 W
Niagara Falls	43 06 N	79 04 W
Nickel Centre	46 34 N	80 49 W
Nipawin	53 22 N	104 00 W
North Battleford	52 47 N	108 17 W
North Bay	46 19 N	79 28 W
North West River	53 32 N	60 08 W
Old Crow	67 34 N	139 50 W
Orillia	44 37 N	79 25 W
Oshawa	43 54 N	78 51 W
Ottawa	45 25 N	75 42 W
Owen Sound	44 34 N	80 56 W
Pangnirtung	66 08 N	65 43 W
Parry Sound	45 21 N	80 02 W
Peace River	56 14 N	117 17 W
Pembroke	45 49 N	77 07 W
Penticton	49 30 N	119 35 W
Percé	48 32 N	64 13 W
Perth	44 54 N	76 15 W
Peterborough	44 18 N	78 19 W
Pictou	45 41 N	62 43 W
Pine Point	60 50 N	114 28 W
Placentia	47 14 N	53 58 W
Port Alberni	49 14 N	124 48 W
Port-Cartier	50 01 N	66 53 W
Port Hawkesbury	45 37 N	61 21 W
Portage la Prairie	49 59 N	98 18 W
Prince Albert	53 12 N	105 46 W
Prince George	53 55 N	122 45 W
Prince Rupert	54 19 N	130 19 W
Quebec	46 49 N	71 14 W
Quesnel	53 00 N	122 30 W
Rae-Edzo	62 50 N	116 03 W
Rankin Inlet	62 49 N	92 05 W
Red Deer	52 16 N	113 48 W
Regina	50 27 N	104 37 W
Resolute Bay	74 41 N	94 54 W
Revelstoke	50 59 N	118 12 W
Rimouski	48 26 N	68 33 W
Rivière-du-Loup	47 50 N	69 32 W
Roberval	48 31 N	72 13 W
Ross River	61 59 N	132 26 W
Rouyn-Noranda	48 14 N	79 01 W
Sachs Harbour	72 00 N	125 13 W
Saint Albert	53 38 N	113 38 W
Saint John	45 16 N	66 03 W
Saint John's	47 34 N	52 43 W
Saint Thomas	42 47 N	81 12 W
Sainte-Foy	46 47 N	71 17 W
Sarnia	42 58 N	82 23 W
Saskatoon	52 07 N	106 38 W
Sault Ste. Marie	46 31 N	84 20 W
Scarborough	43 47 N	79 15 W
Schefferville	54 48 N	66 50 W
Selkirk	50 09 N	96 52 W
Senneterre	48 23 N	77 14 W
Sept-Îles	50 12 N	66 23 W
Shawinigan	46 33 N	72 45 W
Shelburne	43 46 N	65 19 W
Sherbrooke	45 25 N	71 54 W
Sidney	48 39 N	123 24 W
Snow Lake	54 53 N	100 02 W
Springdale	49 30 N	56 04 W
Stony Plain	53 32 N	114 00 W
Sturgeon Falls	46 22 N	79 55 W
Sudbury	46 30 N	81 00 W
Summerside	46 24 N	63 47 W
Surrey	49 06 N	122 47 W
Swan River	52 07 N	101 16 W
Swift Current	50 17 N	107 48 W
Sydney	46 09 N	60 11 W
Teslin	60 10 N	132 43 W
The Pas	53 50 N	101 15 W
Thompson	55 45 N	97 52 W
Thunder Bay	48 24 N	89 19 W
Timmins	48 28 N	81 20 W
Toronto	43 39 N	79 23 W
Trail	49 06 N	117 42 W
Trepassey	46 44 N	53 22 W
Trois-Rivières	46 21 N	72 33 W
Truro	45 22 N	63 16 W
Tuktoyaktuk	69 27 N	133 02 W
Val-d'Or	48 06 N	77 47 W
Vancouver	49 15 N	123 07 W
Vernon	50 16 N	119 16 W
Victoria	48 26 N	123 22 W
Virden	49 51 N	100 56 W
Wabush	52 55 N	66 52 W
Wainwright	52 49 N	110 53 W
Watson Lake	60 04 N	128 42 W
Welland	42 59 N	79 15 W
Wetaskiwin	52 58 N	113 22 W
Weyburn	49 40 N	103 51 W
Whitehorse	60 43 N	135 03 W
Williams Lake	52 08 N	122 09 W
Windsor	42 18 N	83 01 W
Windsor	44 59 N	64 08 W
Winnipeg	49 53 N	97 09 W
Yarmouth	43 50 N	66 07 W
Yellowknife	62 27 N	114 22 W
Yorkton	51 13 N	102 28 W

Physical features and points of interest

Abitibi, *river*	51 03 N	80 55 W
Akimiski Island	53 00 N	81 20 W
Albany, *river*	52 17 N	81 31 W
Amadjuak Lake	65 00 N	71 08 W
Amundsen Gulf	71 00 N	124 00 W
Anticosti Island	49 30 N	63 00 W
Appalachian Mountains	44 00 N	72 00 W
Arctic Ocean	75 00 N	150 00 W
Assiniboine, *river*	51 15 N	101 32 W
Athabasca, *river*	58 40 N	110 50 W
Athabasca, Lake	59 15 N	109 15 W
Atlantic Ocean	45 00 N	50 00 W
Attawapiskat, *river*	52 57 N	82 18 W
Auyuittuq National Park	67 30 N	66 00 W
Avalon Peninsula	47 30 N	53 30 W
Axel Heiberg Island	79 45 N	91 00 W
Back, *river*	67 15 N	95 15 W
Baffin Bay	72 00 N	64 00 W
Baffin Island	68 00 N	75 00 W
Baldy Mountain	51 28 N	100 43 W
Banff National Park	51 30 N	116 15 W
Banks Island	72 45 N	121 30 W
Barbeau Peak	81 55 N	75 02 W
Barnes Ice Cap	70 00 N	73 15 W
Barren Grounds, *region*	62 50 N	105 00 W
Bathurst Island	75 45 N	100 00 W
Beaufort Sea	72 00 N	140 00 W
Belcher Islands	56 15 N	79 15 W
Belle Isle, Strait of	51 35 N	56 15 W
Boothia, Gulf of	71 00 N	91 00 W
Boothia Peninsula	71 00 N	94 00 W
Bow, *river*	49 57 N	111 41 W
Bylot Island	73 13 N	78 34 W
Cabot Strait	47 30 N	59 45 W
Caniapiscau, *river*	57 40 N	69 29 W
Caniapiscau Reservoir	54 10 N	69 50 W
Canso, Strait of	45 37 N	61 22 W
Cape Breton Highlands National Park	46 45 N	60 45 W
Cariboo Mountains	53 00 N	120 15 W
Carleton, Mount	47 23 N	66 53 W
Cassiar Mountains	59 17 N	129 50 W
Caubvick (Iberville), Mount	58 53 N	63 43 W
Champlain, Lake	44 45 N	73 15 W
Chidley, Cape	60 23 N	64 26 W
Churchill, *river*	53 19 N	60 10 W
Churchill, *river*	58 47 N	94 12 W
Claire, Lake	58 35 N	112 05 W
Coast Mountains	54 00 N	129 30 W
Coats Island	62 30 N	83 00 W
Columbia, *river*	49 00 N	117 38 W
Columbia, Mount	52 09 N	117 25 W
Coppermine, *river*	67 49 N	115 04 W
Cumberland Sound	65 10 N	65 30 W
Cypress Hills	49 34 N	110 08 W
Daniel Johnson Dam	50 39 N	68 44 W
Davis Strait	65 00 N	58 00 W
Della Falls	49 27 N	125 32 W
Devon Island	75 47 N	88 00 W
Dixon Entrance	54 25 N	132 00 W
Dubawnt Lake	63 04 N	101 42 W
Ekwan, *river*	53 14 N	82 13 W
Ellef Ringnes Island	78 30 N	103 00 W
Ellesmere Island	81 00 N	80 00 W
Ellesmere Island National Park	83 00 N	73 00 W
Erie, Lake	42 15 N	81 00 W
Evans Strait	63 15 N	82 00 W
Fairweather Mountain	58 54 N	137 31 W
Foxe Basin, *inlet*	68 25 N	77 00 W
Fraser, *river*	49 07 N	123 11 W
Frobisher Bay	62 30 N	66 00 W
Fundy, Bay of	45 00 N	66 00 W
Fundy National Park	45 38 N	65 00 W
Gardiner Dam	51 16 N	106 52 W
Gaspé Peninsula	48 40 N	66 00 W
Georgia, Strait of	49 20 N	124 00 W
Georgian Bay	45 15 N	80 45 W
Glacier National Park	51 16 N	117 31 W
Grasslands National Park	49 11 N	107 38 W
Great Bear Lake	65 50 N	120 45 W
Great Plains	58 00 N	115 00 W
Great Slave Lake	61 23 N	115 38 W

Gros Morne National Park	49 40 N	57 45 W
Hamilton Inlet	54 00 N	57 30 W
Hayes, *river*	57 03 N	92 12 W
Hecate Strait	53 30 N	131 10 W
Henrietta Maria, Cape	55 09 N	82 20 W
Hudson Bay	60 00 N	86 00 W
Hudson Strait	62 00 N	70 00 W
Huron, Lake	44 30 N	82 15 W
Iberville, see Caubvick, Mount		
Jacques-Cartier, Mount	48 59 N	65 57 W
James Bay	53 30 N	80 30 W
Jasper National Park	52 50 N	118 08 W
Jones Sound	76 00 N	86 00 W
Juan de Fuca Strait	48 15 N	124 00 W
Kane Basin, *inlet*	79 01 N	73 00 W
King William Island	69 10 N	97 25 W
Klondike, *river*	64 03 N	139 26 W
Kluane National Park	60 45 N	139 30 W
Kootenay, (Kootenai), *river*	49 19 N	117 39 W
Kootenay National Park	50 40 N	115 55 W
La Grande 2 Reservoir	53 15 N	77 00 W
La Grande 4 Reservoir	53 30 N	72 50 W
Labrador, *region*	54 00 N	62 00 W
Labrador Sea	57 50 N	55 00 W
Lancaster Sound	74 13 N	84 00 W
Liard, *river*	60 00 N	123 48 W
Logan, Mount	60 34 N	140 23 W
Long Range Mountains	50 00 N	57 00 W
M'Clintock Channel	72 00 N	102 00 W
M'Clure Strait	74 30 N	119 00 W
Mackenzie, *river*	69 15 N	134 08 W
Mackenzie Bay	69 20 N	135 45 W
Mackenzie Mountains	64 45 N	132 15 W
Madeleine Islands	47 30 N	61 45 W
Manicouagan Reservoir	51 05 N	68 34 W
Manitoba, Lake	51 00 N	98 45 W
Manitoulin Island	45 50 N	82 20 W
Melville, Lake	53 40 N	59 40 W
Melville Island	75 30 N	111 30 W
Melville Peninsula	68 00 N	84 00 W
Mica Dam	52 03 N	118 36 W
Mistassini, Lake	51 00 N	73 37 W
Monashee Mountains	51 00 N	119 00 W
Moose, *river*	51 20 N	80 24 W
Nahanni National Park	61 35 N	125 45 W
Nechako, *river*	53 55 N	122 42 W
Nelson, *river*	57 04 N	92 30 W
Nettilling Lake	66 30 N	70 40 W
Newfoundland, *island*	48 30 N	56 00 W
Niagara Falls	43 05 N	79 04 W
Nipigon, Lake	49 50 N	88 30 W
North Saskatchewan, *river*	53 15 N	105 05 W
Northern Yukon National Park	69 05 N	139 30 W
Norwegian Bay	77 30 N	90 00 W
Notre-Dame Mountains	48 10 N	68 00 W
Nottaway, *river*	51 22 N	78 55 W
Nouveau-Quebec (Ungava) Crater	61 17 N	73 40 W
Nueltin Lake	60 30 N	99 50 W
Odin, Mount	66 33 N	65 26 W
Ogilvie Mountains	64 35 N	138 15 W
Ontario, Lake	43 45 N	78 00 W
Opinaca Reservoir	52 39 N	76 20 W
Outaouais (Ottawa), *river*	45 20 N	73 55 W
Outardes, *river*	49 04 N	68 28 W
Pacific Ocean	50 00 N	135 00 W
Pacific Rim National Park	49 04 N	125 40 W
Parry, Cape	70 12 N	124 31 W
Parry Channel	74 20 N	98 00 W
Peace, *river*	59 00 N	111 25 W
Peary Channel	79 40 N	101 30 W
Peel, *river*	67 00 N	135 00 W
Pelly, *river*	62 47 N	137 20 W
Pelly Mountains	61 40 N	132 30 W
Penny Ice Cap	67 10 N	66 00 W
Porcupine, *river*	67 25 N	141 00 W
Prince Albert National Park	54 00 N	106 25 W
Prince Alfred, Cape	74 20 N	124 46 W
Prince Charles Island	67 50 N	76 00 W
Prince of Wales Island	72 40 N	99 00 W
Prince Patrick Island	76 45 N	119 30 W
Pukaskwa National Park	48 15 N	85 55 W
Purcell Mountains	50 00 N	116 30 W
Queen Charlotte Islands	53 00 N	132 00 W
Queen Charlotte Sound	51 30 N	129 30 W
Queen Elizabeth Islands	78 00 N	95 00 W
Queen Maud Gulf	68 25 N	102 30 W
Red, *river*	50 24 N	96 48 W
Red Deer, *river*	52 55 N	101 30 W
Reindeer Lake	57 15 N	102 40 W
Richardson Mountains	67 00 N	136 10 W
Riding Mountain National Park	50 55 N	100 25 W
Robson, Mount	53 07 N	119 09 W
Rocky Mountains	54 30 N	122 30 W
Ronge, Lake	55 10 N	105 00 W
Rupert, *river*	51 30 N	78 45 W
Sable Island	43 57 N	59 55 W
Saint Clair, Lake	42 28 N	82 40 W
Saint Croix, *river*	45 05 N	67 06 W
Saint Elias Mountains	60 33 N	139 28 W
Saint-Jean, Lake	48 35 N	72 05 W
Saint Lawrence, *river*	45 12 N	74 20 W
Saint Lawrence, Gulf of	48 00 N	62 00 W
Saskatchewan, *river*	53 30 N	100 15 W
Sault Ste. Marie Bridge	46 30 N	84 20 W
Selkirk Mountains	50 30 N	117 30 W
Selwyn Mountains	63 00 N	130 00 W
Severn, *river*	56 02 N	87 36 W
Skeena, *river*	54 01 N	130 07 W
Slave, *river*	60 09 N	111 49 W
Smallwood Reservoir	54 05 N	64 30 W
Smith Sound	78 25 N	74 00 W
Somerset Island	73 15 N	93 30 W
Souris, *river*	49 00 N	102 00 W
South Moresby National Park	50 35 N	131 50 W
South Saskatchewan, *river*	53 15 N	105 05 W
Southampton Island	64 20 N	84 40 W
Stikine, *river*	56 39 N	131 50 W
Superior, Lake	48 00 N	87 00 W
Takakkaw Falls	51 30 N	116 28 W
Terra Nova National Park	48 33 N	53 55 W
Thelon, *river*	64 16 N	96 05 W
Torngat Mountains	59 00 N	63 40 W
Trinity Bay	48 00 N	53 30 W
Ungava, see Nouveau-Québec Crater		
Ungava Bay	59 30 N	67 30 W
Ungava Peninsula	60 00 N	74 00 W
Vancouver Island	49 30 N	125 30 W
Victoria Island	71 00 N	110 00 W
Virginia Falls	61 36 N	125 44 W
Viscount Melville Sound	74 10 N	108 00 W
W.A.C. Bennett Dam	56 01 N	122 12 W
Waterton Lakes National Park	49 05 N	113 52 W
Williston Lake	56 00 N	124 00 W
Winisk, *river*	55 17 N	85 05 W
Winnipeg, Lake	52 00 N	97 00 W
Winnipegosis, Lake	52 30 N	100 00 W
Wollaston Lake	58 15 N	103 20 W
Wood Buffalo National Park	59 15 N	113 15 W
Woods, Lake of the	49 15 N	94 45 W
Yellowhead Pass	53 53 N	118 28 W
Yoho National Park	51 30 N	116 30 W
Yukon, *river*	64 41 N	141 00 W

Temperatures. Those parts of the country farthest from open water are the coldest, so that in the interior plains and in the North the winters are extremely cold. The lowest temperature ever recorded was −81° F (−63° C) at Snag in the Yukon. In July those parts of Canada farthest from open water are the warmest. Thus, west-coast Vancouver has an average January temperature of 37° F (3° C) and an average July temperature of 64° F (18° C), while on the interior plains Regina's average temperatures vary from −1° to 67° F (−18° to 19° C). The daily range of temperature is also less on the coasts than in interior locations. The highest temperature recorded was 115° F (46° C) at Gleichen in southern Alberta.

Impact of the Pacific and Rockies on precipitation

Rainfall. Humid air masses from the Pacific cause enormous quantities of orographic (mountain-caused) rain to fall on the coast and mountain areas. Several sites along the British Columbia coast receive annual precipitation in excess of 100 inches (2,540 millimetres). But British Columbia also receives much less precipitation in summer than in winter because low-pressure systems move on a more northerly track in summer and seldom cross the southern part of the coast. Vancouver has an annual average precipitation of 41 inches.

In the interior plains and the North (Arctic and subarctic), precipitation is seldom more than 15 inches a year; it drops as low as 2 inches at Eureka on Ellesmere Island. The west-coast mountains effectively keep marine air out, since the dominant movement of air is from west to east. Spring and summer are wetter than winter.

Ontario and Quebec have more rainfall than the interior plains because the air masses can pick up water vapour from the Great Lakes, Hudson Bay, the Atlantic Ocean, and the Gulf of Mexico. The average annual precipitation for Toronto is 31 inches (800 millimetres); for Montreal, 40 inches (1,020 millimetres). Because winters are not as cold as those in the interior plains, the air is less dry, and enough snow falls to make winter precipitation equal to that in summer. The Atlantic Provinces are wetter than those of Central Canada. Yearly rainfall, most of which is cyclonic in origin, is more than 50 inches in places and is fairly evenly distributed throughout the year. There are few thunderstorms, and the low Appalachian Mountains produce only a little orographic rainfall. In general, the rainfall on Canada's east coast is less than on the west coast because the prevailing wind is offshore.

Snowfall. Canada's snowfall does not follow the same pattern as rainfall. In the North and the interior plains, snowfall is light because cold air is very dry. The snow is hard and dry, falls in small amounts, and is packed down by the constant wind. The east and west coasts are areas of lighter snowfall because the ocean usually makes the air too warm for large quantities of snow to fall. The depth of snow increases inland from each coast, reaching maximums of about 240 inches in the Rocky Mountains and on the shores of the Gulf of St. Lawrence. Still farther inland, lack of moisture brings the depth of snow down again. Freezing precipitation may occur during the colder months in any part of the country, occasionally disrupting transportation and communication.

PLANT AND ANIMAL LIFE AND SOILS

Both landforms and climate affect the distribution of plants, animals, and soils. Ecologists recognize broad regions called ecosystems that are characterized by fairly stable complexes of climate, soils, and plant and animal life. The boundaries of these regions are not usually sharp lines on the landscape, but are broad transition zones. The discussion that follows concentrates on preagricultural, or natural, vegetation. In southern Canada only remnants of these ecosystems remain.

Vulnerability of tundra

Tundra. Tundra is the land type of the Arctic and subarctic regions. Tundra also exists above the tree line in the Western Cordillera, but the discussion here is generally confined to the northern tundra. With long cold winters, short cool summers, and low precipitation, the soils are thin or absent, and the vegetation is sparse. The tundra is the ecosystem most susceptible to environmental damage. Because of the small number of plant and animal species and the fragility of the food chains, damage to any element of the habitat may have an immediate chain reaction through the system. The permafrost is easily damaged by heavy equipment and by oil spills. The Inuit, who fish, hunt, and trap for a living, are directly affected by abuses of the ecology.

Tundra vegetation is quite varied, considering the climatic conditions. Although the rock deserts are almost devoid of vegetation, relatively fast-growing mosses often surround large rocks. In rock crevices such plants as the purple saxifrage survive, and the rock surfaces themselves may support lichens, some of the orange and vermilion species adding colour to the landscape. Lichen tundra is found in the drier and better-drained parts. Mosses are common, and some species may dominate the landscape to such an extent that it appears snow-covered. The heath and alpine tundra support dwarf, often berry-bearing shrubs, while the ground between usually is covered with a thick carpet of lichens and mosses.

The distinctive animals of the tundra are the seal and the polar bear, the latter feeding on seal; the musk-ox, caribou, and lemming, which feed on the tundra vegetation; and their predators, the Arctic wolf and white fox. Few birds make the tundra their year-round habitat, the great snowy owl and the ptarmigan being exceptions. Numerous birds that normally live in mild climates, however, often fly to the tundra for nesting. Two large birds that do this are the snow goose and the Canada goose.

The forest regions. Canada has several large and distinct forest zones, which blend into a number of transitional zones. The northern coniferous, or boreal, forest is the world's second largest area of uninterrupted forest; only the Soviet Union has a greater expanse of forest. With severe winters and a short growing season, there is a limited number of tree species. Among them the white and black spruce and white birch are common, and balsam, fir, and tamarack (larch) also have wide distribution. The boreal forest is an important source of pulpwood and also produces considerable lumber, but much of the northern area is too inaccessible for commercial lumbering.

Transitional zones

A vast transitional zone of mixed boreal and tundra growth connects the northern forest and the tundra region. The trees in this subarctic region, with its cold, dry climate, are small and of little commercial consequence. Underlaid with intermittent permafrost, its ecological balance is almost as delicate as that of the tundra.

Along the southern edge of the coniferous forest lie two other transitional zones. In the interior plains the forest merges with the grasslands to create an arc of aspen parkland, characterized by prairie vegetation dotted with groves of trembling aspen and other poplar species in low moist areas and along valley bottoms. East of the Manitoba–Ontario border is a band of mixed coniferous-deciduous forest that extends into the Great Lakes–St. Lawrence lowlands and into the Appalachian region. In addition to the species of the boreal forest, there are white pine, red pine, white cedar, and eastern hemlock. The deciduous trees include sugar maple, red maple, beech, red oak, and white ash.

Remnants of the only predominantly deciduous forest in Canada grow in the most southerly portion of the southwestern Ontario peninsula. This deciduous forest is an extension of the Carolinian forest zone of the United States, and in addition to the species it shares with the mixed forest, it contains species usually found much farther south, such as the tulip tree, sycamore, black and white oak, and several types of hickory.

The western forests

The forests of the Western Cordillera are complex, as might be expected from the strong relief and the sudden change in climate within relatively short distances. The subalpine forest, of Engelmann and white spruce and lodgepole pine, is characteristic of the slopes of the Rockies from about 4,000 feet up to the timberline. The forests of the Selkirk, Purcell, and Monashee mountains contain Engelmann spruce at higher elevations, merging with western red cedar and western hemlock on the lower slopes. On drier slopes Douglas fir is common. Typical of the ranges that traverse the rather arid interior plateau is a generally open forest of aspen and yellow pine interspersed with glades of grass. Douglas fir and lodgepole pine are found on higher slopes.

Pacific coastal forest

The forest of the Pacific coast, where steep slopes facing rain-bearing winds produce a high rainfall, is Canada's densest tall-timber forest. Abundant moisture, along with the long growing season, is conducive to the growth of evergreens with very hard wood, excellent for construction lumber. Douglas fir, western hemlock, and western red cedar are the outstanding trees; they grow to great height and thickness. Alder, cottonwood, and maple are subsidiary, along with western white pine. Dense stands of immense trees, their trunks rising to considerable heights, their crowns almost touching, give a grandeur to the forest.

Canada's forest soils are acidic, resulting in various degrees of leaching of minerals out of the topsoil, thus making them relatively infertile for agriculture. The degree of acidity and leaching is greater in the coniferous forest and less in the mixed and deciduous forest. With proper soil-management practices, the mixed and deciduous forest soils make good farmland.

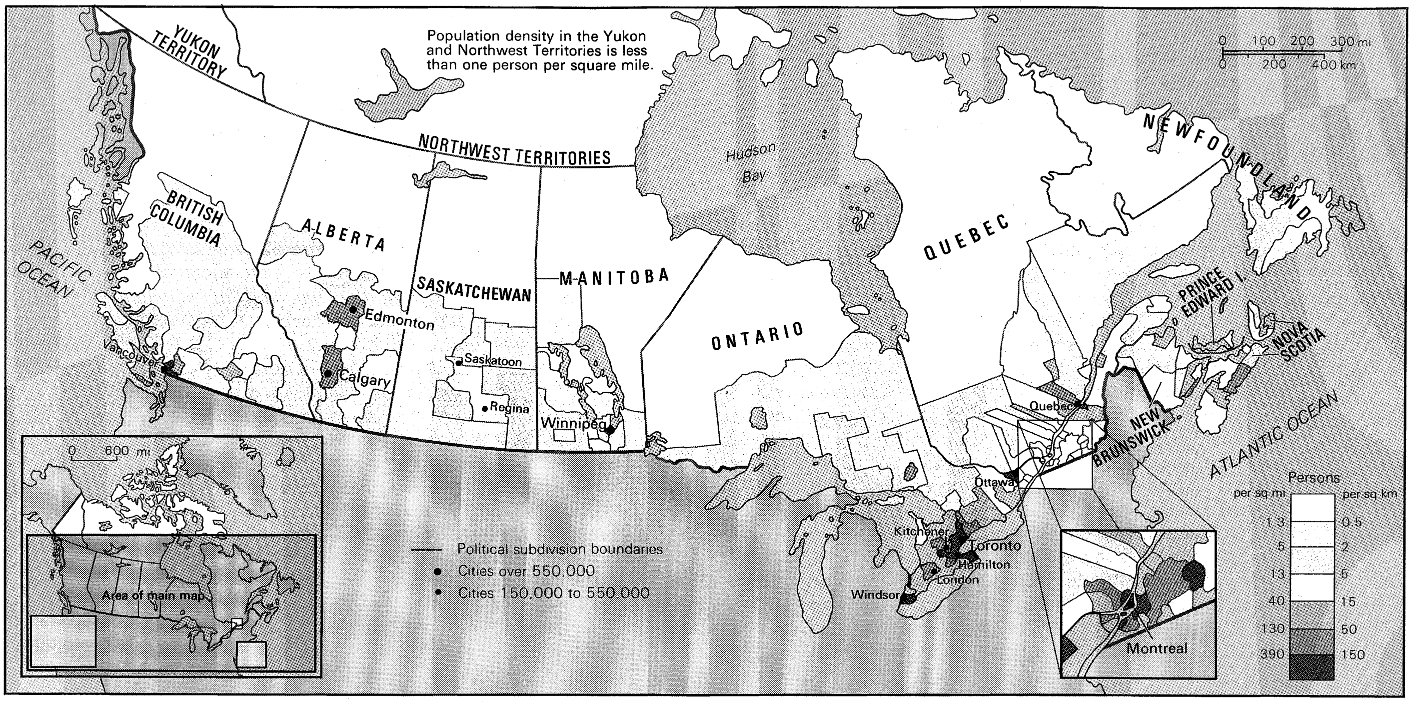
Wildlife regions correspond closely to the different forest zones. The subarctic supports large numbers of woodland caribou. The boreal forest includes nearly all species of mammals and birds recognized as distinctively Canadian. Among these are the moose, beaver, Canada lynx, black bear, wolf, snowshoe hare, and a variety of birds, including the Canada jay, blue jay, gray jay, raven, and crow. In summer the coniferous forest fills with scores of varieties of warblers and other small birds that go north to nest. Farther south the white-tailed deer thrives on the forest borders and partially cleared areas. There also are numerous smaller mammals, including the gray and red squirrels, mink, raccoon, muskrat, skunk, jackrabbit, cottontail rabbit, groundhog, and a variety of mice and moles. In southern Ontario the wild turkey, which disappeared because of hunting and reduction of its habitat, was reintroduced in the 1980s with some success. A wide range of wildlife species inhabit the Western Cordillera, with its wide variety of terrain and vegetation. Rocky Mountain sheep, mountain goats, elk, mule deer, and black bears are common in the southern mountains.

Grasslands. The southern portion of the interior plains is too dry for forests and gives rise to grasslands or natural prairies. Shortgrass with sagebrush and cactus make up the natural vegetation of the most southerly area. Farther north, where there is slightly more precipitation, there is a band of tallgrass prairie. At its northern limit the grasslands merge with the transitional parkland at the edge of the boreal forest. Today the grass area is small, crops having replaced grass in all but dry or hilly areas.

Fertile grasslands

With its high organic matter and mineral content, the grassland soils are among the most fertile in Canada. The best soils for crops are the dark brown to black soils of the tallgrass and parkland zone, the area of Canada that has become famous for wheat growing. The less fertile light brown soils of the shortgrass country tend to be very alkaline, and the predominant agricultural activities are dryland farming and grazing. Wind erosion is a serious problem in the prairies wherever the grassland has been converted to cultivated farmland.

Among the common animals of the grasslands are Richardson's ground squirrel and the pocket gopher, both of which damage young grain crops. They continue to proliferate despite predation by badgers, hawks, and owls and attempts at control by farmers. The first settlers to cross the Canadian prairies encountered enormous herds of bison (often called buffalo), but by the beginning of



Population density of Canada

the 20th century hunters had reduced their numbers to near extinction. Bison may now be seen only in wildlife reserves. With the bison gone, the mule deer and the pronghorn antelope are the remaining large animals on the shortgrass plain. Farm drainage projects and extended drought have greatly reduced the prairie's waterfowl habitat, causing a decline in their numbers.

SETTLEMENT PATTERNS

When the Europeans began exploring and developing resources in what is now Canada, they found the land sparsely populated by many different Indian peoples in the south and the Inuit in the north. The native peoples were primarily hunters and gatherers and often were nomadic. Because they were few in number, the native peoples made little impact on the natural environment; they harvested only the resources needed for their own consumption, and there were no large settlements. Even though the native peoples had lived in the area for thousands of years, the Europeans perceived that they had found a pristine country with rich resources that awaited exploitation.

Different groups of Europeans came at different times to develop and export the natural resources of fish, furs, forests, and minerals. With the development of each new resource, new settlements were established. Most of the settlements based on these resources remained small, however, and some of them disappeared when the resource was depleted. A few port cities—including the eastern cities of St. John's, Nfd.; Halifax, N.S.; and Saint John, N.B.—continued to grow, as they benefited from the export of successive resources. Montreal owed its early growth to the fur trade, but later it became an important entrepôt for the export of a succession of raw and processed materials and the import of manufactured goods from Europe. Later, Toronto and, on the west coast, Vancouver also grew quickly because of entrepôt activities. Winnipeg, Man., owed its early growth to its gateway role in the agricultural development of the interior plains.

Except for the port cities, Canada's most densely settled areas and largest cities have developed in the areas with good agricultural land. Thus, within a narrow strip of land along the U.S.–Canada border, making up about 10 percent of the total national area, live some 90 percent of the people. Intensive commercial agriculture in the Great Lakes–St. Lawrence lowlands gave rise to a dense network of villages, towns, and cities. Later, manufacturing and service industries reinforced population growth in this region, making it the urban and industrial heartland of Canada. Villages, towns, and cities also evolved from the agricultural industry in the western grasslands, but due to a lack of growth in the manufacturing and service industries, much less urbanization occurred. The development of the petroleum industry, however, did stimulate the growth of two large cities, Edmonton and Calgary in Alberta.

At the beginning of the 20th century, only one-third of the people of Canada lived in urban areas, but by the late 20th century about four-fifths of the population lived in communities of more than 1,000 people and more than one-half were in cities of 100,000 or more.

The growth of most of Canada's large cities on good farmland with an accompanying low-density pattern of urban sprawl has aroused considerable public concern about the urbanization of Canada's limited agricultural

George Hunter

Long, narrow strips of farmland, typical of French Canada, laid out along the roads in the St. Lawrence River valley, near Montreal, Que.

land resources. In the Niagara Peninsula of southwestern Ontario, an area with the best climate in Canada for the production of soft fruits and grapes, urbanization has destroyed about one-third of the fruit land. To prevent further reduction, the Ontario Municipal Board in 1981 delineated permanent urban boundaries and ordered that urban growth be directed away from fruit-growing areas.

Cultural diversity of settlement

Settlement did not proceed westward from an Atlantic beginning because permanent settlement was dependent upon agricultural land, and agricultural land in Canada occurs in patches, separated by physical barriers. Different patches were settled by people from various European countries, resulting in the diversity of cultures and settlement patterns across the country.

In the Appalachian region, farms are spaced along the roads at irregular intervals wherever there is cultivable land. In Quebec the first settlers laid off long, narrow lots from the shores of the Gulf of St. Lawrence or the St. Lawrence River into the interior. As settlement moved farther inland, roads were built parallel to the waterways, from which narrow lots extended on either side. The same pattern is seen in the Red River valley of Manitoba, where the early settlers were also French.

In Ontario and the eastern townships of Quebec, land subdivision was made with reference to a much more definite plan. The townships were more or less square, but the grid became irregular because it was started from a number of different points, each of which used a differently oriented base. In the prairies of the interior plains the grid is much more regular, partly as a result of the topography, partly because a plan for the subdivision of the whole region was laid out in advance of settlement.

The people

ETHNIC COMPOSITION

Canadians do not form a compact, homogeneous people. They are, rather, a collection of diverse national and cultural groups. In the strictly legal sense, there was no such thing as a Canadian citizen until the Canadian Citizenship Act came into force on Jan. 1, 1947.

Principal ethnic groups. About one-half of Canadians descend from one of the two founding nations, Britain or France. At the time of the first census of Canada (1871), about 50 percent of the population was British and 30 percent French. Since that time the proportion of Canadians of British and French origin has dropped to about 25 percent each. This decline has resulted from a reduction in the number of people coming from the United Kingdom and France and an increase in the number from other countries in Europe, Southeast Asia, and Latin America. Immigrant groups tend to retain their cultural identity in Canada largely because they have tended to settle in colonies. For example, Ukrainians have been attracted to the prairies, where the land and climate are similar to their homeland, and many Dutch have settled on the flat fertile farmland of southwestern Ontario, where they practice fruit and vegetable growing as they did in The Netherlands. Many Chinese, Portuguese, Greeks, and Italians have settled in specific sections of large cities, particularly Toronto, Montreal, and Vancouver.

Retention of cultural identity

Census data show that the mix of ethnic origins differs greatly from province to province. The proportion of people of British origin ranges from about 80 percent in Newfoundland to 5 percent in Quebec; the proportion of people of French descent ranges from about 80 percent in Quebec to less than 3 percent in Newfoundland, British Columbia, and the territories.

U.S. immigration. Canadian census data do not account for the number of Canadians whose forebears emigrated from the United States. There was heavy migration from the United States during and after the American Revolution (1776–83), when colonists who remained loyal to the British crown (known as United Empire Loyalists in Canada) moved to what is now the Maritime Provinces and southern Quebec and Ontario. By 1790 about one out of every six people in British North America was of U.S. origin. These people had been exposed to the ideas of the popular government and democracy that had evolved along the Atlantic seaboard. Their ideas of governmental institutions were blended in Canada with those of people who came directly from Britain. There was some migration from the United States to Canada during the 19th and 20th centuries, but during the same period there was more movement of Canadians to the United States, due to the generally higher U.S. standard of living.

Native peoples. It is estimated that, when Europeans began settling what is now Canada, there were only about 200,000 Indians and Inuit. For about 200 years after the first permanent European settlement, the native population declined somewhat because of European encroachment on their territory. High birth rates and improved medical care in the last half of the 20th century, however, have greatly increased the native population.

American Indian reserves

In Canada the word Indian has a legal definition given in the Indian Act of 1876. People legally defined as Indians are known as status-Indians. Indians who have chosen to give up their status rights or who have lost them through intermarriage with whites are called nonstatus Indians. Through treaties with the Canadian government, some 542 status-Indian bands occupy more than 2,250 reserves. The resources of these reserves are limited, and the majority of status-Indians have a standard of living below the Canadian average. The treaties and agreements about reserves apply only to a portion of the Indian people. Large tracts of land were never taken from the Indians by treaty, and various groups are negotiating land claims with the Canadian government.

The Inuit who inhabit the far north do not have any reserves and are not covered by any treaties. Their estimated population of more than 27,000 live in scattered camps and settlements of 25 to 500 people. In the latter part of the 20th century, mining, oil exploration and pipeline construction, and mammoth hydroelectric developments greatly affected their traditional way of living off the land. The worldwide decline in demand for furs greatly diminished their income, and they became increasingly dependent upon government social and welfare programs. Education and training programs were instituted to enable them to compete for employment.

LINGUISTIC COMPOSITION

The 1986 census of Canada reported that about 60 percent of the people identified English as their first language, while about 24 percent said that French was their mother tongue. Some 11 percent of Canadians reported a first language other than English or French; most of these indicated that they spoke another European language, while a smattering reported Asian and Middle Eastern tongues as their first language. English and French are both official languages of Canada. The language of the Inuit, called Inukitut or Inuttituut, has a number of variations.

RELIGIONS

More than 85 percent of Canada's population claim affiliation in some degree with either the Roman Catholic or a Protestant church. The major Protestant churches are the United Church of Canada, the Anglican Church of Canada, and the Lutheran Church. The balance of the population adhere to the Jewish, Eastern Orthodox, and other faiths or have no affiliation. In Quebec more than 80 percent of the population is Roman Catholic, and New Brunswick also has a Roman Catholic majority. The other provinces have Protestant majorities.

DEMOGRAPHIC TRENDS

Immigration

Traditionally Canada has sought to increase its population through immigration in order to expand the work force and domestic markets. The peak year of immigration in the 20th century was 1913, with more than 400,000 arrivals. Immigration was discouraged during the Great Depression of the 1930s, but after World War II tens of thousands of displaced persons from Europe were admitted, and in the 1970s and '80s large numbers of refugees from Europe, Asia, and Latin America were welcomed to Canada. Canada's immigration policy is nondiscriminatory regarding ethnicity; however, individuals with special talents or with capital to invest are given preference.

During the first two decades of the 20th century, the notable feature of internal migration was the movement from eastern Canada to the Prairie Provinces. Although British Columbia has continued to gain from migration since 1931, much of this has been at the expense of the Prairie Provinces. Alberta gained population from all across Canada during the oil boom of the 1970s. Saskatchewan has had more emigration than immigration since 1941. Ontario consistently has received far more people since 1941 than the other provinces, but most of this growth has been from immigration rather than interprovincial migrations. The Atlantic Provinces have lost about 300,000 people since World War II.

During the 20th century, natural increase, not immigration, was the major factor in Canada's population growth. Until 1961 the crude birth rate (live births per 1,000 population) remained in the high 20s, while the crude death rate (deaths per 1,000 population) declined from more than 10.6 in 1921 to 7.7 in 1961. The rate of natural increase slowed after 1961, however, because of a sharp drop in the birth rate accompanied by a slight decrease in the death rate. The rate of natural increase is much lower than the world average and is about the same as that of the United States and Australia. Canada has an aging population. The life expectancy is longer for Canadians than it is for the peoples of most other countries. The percentage of the Canadian population over 65 years of age more than doubled in the 20th century.

The economy

The early settlement and growth of Canada depended on developing and exporting natural resources. During the 20th century, manufacturing and service industries became increasingly important. By the end of the 1980s agriculture and mining accounted for less than one-tenth of Canada's labour force, while manufacturing stood at one-fifth and the service industry, including transportation, trade, finance, and other service industries, at about two-thirds. For many years Canada encouraged its manufacturing industries through protective tariffs on the import of manufactured goods. This high tariff policy resulted in the establishment of Canadian branch plants by many U.S. firms in order to supply the Canadian market. Another cornerstone of Canada's economic policy was the encouragement of economic development in slow-growth regions of the country by grants and subsidies. In the 1980s Canada began moving away from these two basic policies. Compliance with international rules on trade and the signing of a free-trade agreement with the United States (1988) reduced protection for the manufacturing industry. Funding for regional economic development programs was also scaled down.

Private-sector orientation

Canada's economy is oriented to the private sector, with only relatively few enterprises such as postal services, electric utilities, and some transportation services publicly owned. The agricultural industry is firmly private, although it has had to become dependent on government subsidies in order to compete with the highly subsidized agricultural industries of the European Economic Community (EEC) and the United States. Several marketing boards for specific farm commodities practice supply management and establish floor prices.

About one-third of Canada's labour force belongs to trade unions. Of this total, two-thirds belong to unions with international affiliations, mostly with unions in the United States. The Canadian unions tend to strive for wage parity with their counterparts in the United States. This causes labour–management tensions because Canadian productivity levels are generally lower than those in the United States, due primarily to smaller production runs.

RESOURCES

Minerals. Canada is rich in mineral resources. The vast Canadian Shield, with its masses of igneous and metamorphic rocks, contains numerous large deposits of iron ore and such other metallic minerals as gold, silver, copper, platinum, nickel, lead, zinc, titanium, cadmium, molybdenum, uranium, and cobalt. Metallic minerals are also found in the igneous and metamorphic zones of the Western Cordillera and the Appalachians. Although there are some metallic mineral and fossil fuel deposits in sedimentary rocks in the Western Cordillera and the Appalachians, the largest volume of coal and petroleum has so far been found in the interior plains of western Canada.

Farmland. According to the Canada Land Inventory only 8 percent of Canada's land area is suitable for crop production. About 80 percent of this cropland is in the western prairies, where long sunny days in summer and adequate precipitation combine to provide good yields of healthy grain crops. The widest range of crops, however, and the highest yields occur in southwestern British Columbia and southern Ontario.

Forests. Almost one-half of Canada's land area is covered with forest. The accessible portions of the huge coniferous forest zone provide abundant resources for lumber and pulp and paper. The most valuable forest region for timber production is the west coast, where the climate is conducive to the growth of giant trees with excellent-quality lumber.

Fish. Canada has rich fishing grounds off both the Atlantic and Pacific coasts. In 1977 Canada established a 200-mile fishing zone along both coasts to give the country more control over the management of its ocean fish resources. Off the Atlantic coast the continental shelf extends seaward for several hundred miles. The parts of the continental shelf with the shallowest water are known as fishing banks, where plankton, on which fish feed, thrive because the sunlight penetrates to the seafloor. The most important of these fishing banks is the Grand Banks of Newfoundland. Off the coast of the Maritime Provinces are found Bradelle Bank, Sable Bank, Georges Bank (shared with the United States), and a number of other fishing banks. On the Pacific coast the continental shelf is very narrow, but the numerous mountain streams are suitable for salmon spawning. In the rivers of the far north and in the Arctic Ocean there are abundant fish upon which large numbers of the native peoples depend for food. Fish resources in southern Canada declined significantly after the mid-20th century because of overfishing and pollution.

Grand Banks fishing grounds

Energy. Canada is richly endowed with hydroelectric power resources. It has about 15 percent of the world's total installed hydroelectric generating capacity. However, most of the suitable hydroelectric sites have already been developed, except in Quebec, which has huge hydroelectric potential in the James Bay area. Increasingly, Canada is turning to coal-fueled thermal energy or nuclear energy. Canada has vast coal reserves, particularly in the western provinces, except Manitoba, and in New Brunswick and Nova Scotia. Canada can almost meet its own petroleum needs and has a surplus of natural gas. The largest producing oil and gas fields are in Alberta, but potential reserves lie both in the Arctic and off the east coast. There are also large deposits of uranium.

AGRICULTURE, FORESTRY, AND FISHING

Agriculture. Although agriculture employs only about 5 percent of the Canadian labour force, it is a very important Canadian industry, producing large volumes of food for both the domestic and export markets and providing raw materials for the food processing industry and the food wholesale and retail industry. The trends in the industry are toward larger farm units, mechanized farm operations, specialization in fewer products, and the use of improved varieties, breeds, and farming methods.

There are distinctive types of farming across the country. The Prairies are known for grain (particularly wheat), oilseeds, and cattle grazing. Central and eastern Canada have a wider variety of crops and livestock, and farmers tend to specialize in either a particular cash crop or livestock type. Southwestern Ontario produces large amounts of corn (maize) for grain and soy and white field beans. Both southwestern Ontario and southwestern British Columbia produce a wide variety of fruits and vegetables. Dairying is important around all the major cities.

Forestry. Forest products form a larger part of Canada's export trade than do the combined exports of farm, fish, and mineral products. Canada is the world leader in the export of pulp and paper and also exports large amounts of lumber, mostly to the United States. British Columbia, Ontario, and Quebec are the leading provinces in the production of forest products. Canada's forest industry has to struggle constantly against the threats of fire, insects, and disease. Some control of insects has been afforded by aerial spraying of insecticides, but this practice also eradicates insect-eating birds and predator insects. Because forests have significant value in maintaining an ecological balance in the environment and also provide important recreational opportunities, the forest industry is increasingly held to account for environmentally damaging practices.

Environmental problems in forestry

Fishing. Canada catches only a very small percentage of the fish taken from the world's oceans, but it ranks first among countries in volume of fish exports. The exports are high because of Canada's relatively small population and quite low per capita fish consumption. In the Atlantic Provinces the fishing industry contributes significantly to the value of all goods produced in the region. Small coastal communities throughout the region are wholly or partly dependent on the fishing industry. The most important species caught on the east coast are cod, haddock, redfish, flatfish, turbot, pollock, flounder, sole, halibut, herring, mackerel, tuna, and lobster. In the Pacific coast fishery, salmon is the leading species caught, followed by herring, the two constituting most of the catch. On both coasts the Canadian government has established catch quotas for species and has negotiated international agreements toward preserving fish stocks and protecting the Canadian fishing industry.

INDUSTRY

Mining. Canada is one of the leading mineral producers in the world. In the late 1980s it ranked first or second in the world as a producer of uranium, zinc, nickel, potash, asbestos, gypsum, elemental sulfur, and titanium. It is also a major producer of iron ore, coal, petroleum, gold, copper, silver, lead, and a number of ferroalloys. Modern mining is not labour-intensive, so the industry employs a very small portion of the Canadian labour force. Related industries such as iron and steel and transportation, however, account for a much larger share of the labour force. Because Canada exports a large proportion of its mineral production, the mining industry is sensitive to world price fluctuations. In times of high world demand, prices rise and mining companies step up their production and open new mines; when the demand falls, production is cut, mines close, and workers are laid off. In single-industry communities, mine closings can result in ghost towns. Mining has been a key factor in developing Canada's northlands. In many areas, roads and railroads built to serve new mining operations have encouraged the subsequent development of forest and recreational resources. Unfortunately, development is often accompanied by environmental damage.

Manufacturing. Manufacturing accounts for about one-fifth of Canada's gross national product and labour force. Canada's iron and steel industry is modern and efficient and produces steel products for the manufacture of such durable goods as automobiles, mining equipment, and household appliances. After the Canada–U.S. Automotive Products Agreement, signed in 1965, the Canadian automobile industry expanded dramatically; automobile manufacturing was largely in branch plants of U.S. firms until the 1980s, when Japanese firms began building automobile factories in Canada. Also in the 1980s, the high technology and electronics industries experienced rapid growth. Although there is some manufacturing in all large cities, more than three-fourths of Canadian manufacturing employment is located in the Canadian heartland, which extends from Quebec city to Windsor, at the edge of the Detroit automobile-manufacturing centre.

Canadian automobile industry

Service industries. Service industries in Canada employ more people than all other industries combined. Among the fastest growing of Canada's service industries is tourism. Canadian and foreign travelers spend several billion dollars each year on transportation, accommodations, food, recreation, and entertainment as they travel in the country. By 1990 the tourist industry was providing employment for about 5 percent of Canada's total labour force.

FINANCE

Canadian financial services have developed relatively flexibly in response to the monetary needs of the economy. To operate in Canada, a commercial bank must be individually chartered by Parliament. Most normal central-banking functions are fulfilled by the Bank of Canada, which has substantial autonomy in the determination of monetary policy. The Bank of Canada implements its monetary policies through its relations with the 11 chartered (commercial) banks, which are at the centre of the financial system and rather highly developed. Although there are other financial institutions—credit unions, provincial savings banks, trust and mortgage-loan companies—the chartered banks remain the main financial institutions, even in some specialized activities. They are relatively free from controls on activities involving foreign exchange.

Canada has five stock exchanges, in Alberta, Montreal, Toronto, Vancouver, and Winnipeg. There is extensive interpenetration between Canadian and U.S. stock exchanges. In the bond market the role of government-sector borrowing traditionally has been dominant. The degree of foreign ownership of Canadian industry is very high, as much as 50 percent in the primary resource industries (except agriculture) and manufacturing. The largest portion of the foreign investment is from the United States.

TRADE

Canada's history has always been strongly associated with trading. In the early days, Canada's economic development was dependent upon the export of large volumes of raw materials. Over the years, raw materials have declined as a percentage of Canada's exports, while processed, fabricated, and manufactured goods have increased. By 1990 approximately 80 percent of Canada's exports were processed to some degree. Since about the mid-1970s, the leading Canadian exports have been automobiles and automobile parts, and other types of machinery and equipment, including such high-technology products as portable electrical transformers, computerized communication systems, and nuclear reactors. Fabricated metals and other materials and forestry products, including wood pulp and newsprint, are other important exports.

Canadian exports

The largest proportion of imports in Canadian trading history has always fallen within the manufacturing category. Automobiles and automobile parts lead the list of imports, followed by industrial machinery. Other significant imports are chemical products, textiles and clothing, and such foods as vegetables in the winter season and tropical and subtropical fruits and nuts.

Because the United States is so close and has a common language and similar life-style to that of the majority of Canadians, that country is Canada's chief trading partner. The 1988 free-trade pact between Canada and the United States further enhanced the volume of trade between the two countries. In the late 1980s some 75 percent of Canada's foreign trade was with the United States.

The dependence on U.S. trade in particular is not just a technical matter of market shares in imports and exports. In an overall sense, the degree of export dependence means that business trends in the United States feed back directly and quickly into the Canadian business sector. Changes in consumer tastes in the United States may have disproportionate effects on Canadian producers because to North American consumers it is a matter of almost no relevance whether, for example, their automobiles are produced in the United States or Canada.

Canada also retains strong ties with EEC and other traditional European trading nations, but newly emerging trading patterns may decrease somewhat Canada's dependence on its traditional trade partners. Foremost among these new patterns is that which has developed within the Pacific Rim. The growing Japanese market resulted in Japan's replacing the United Kingdom as the second largest market for Canadian exports by the mid-1970s. Since that time, exports to China, South Korea, and Australia have increased.

TRANSPORTATION

An efficient transportation system is essential to Canada because of its great size, the patchiness of its population distribution, and the need to move primary and manufactured goods over long distances to coastal ports.

Roads and highways. The populated sections of Canada are well traversed by highways and roads, but vast areas of the larger provinces and the territories that are sparsely settled are virtually without roads of any kind. Access to outlying settlements often is provided by roads built by logging, pulp and paper, and mining companies, although these are not generally available for public travel. The Trans-Canada Highway was opened officially in 1962, and it became possible to drive the entire length of the 4,860-mile route from St. John's, Nfd., to Victoria, B.C. Ferry connections extend the highway on both coasts. In the urban-industrial heartland, a four-lane highway links Windsor to Quebec city. Motor vehicles number more than one for every two inhabitants. Trucking as an industry has grown spectacularly since World War II.

The Trans-Canada Highway

Railways. Canada has one of the highest railway mileages per capita of any country in the world. The main railway networks, although connecting the Atlantic and Pacific coasts, are confined to the southern part of the country. Even in the west, where they extend farthest north, the transcontinental routes do not go farther north than Edmonton and Prince Rupert. North–south regional lines, however, reach Hudson Bay at Churchill, Man.; James Bay at Moosonee, Ont.; and central Labrador at Schefferville, Que.

Two great transcontinental systems operate almost all the railway facilities in Canada. The Canadian National Railways (CN) system, a government-owned body, is the country's largest public utility. The Canadian Pacific Railway Company (CP) is a joint-stock corporation. These systems, though highly competitive, still cooperate in many fields in which duplication of service would not be profitable. They are supplemented by a major north–south line on the west coast, the British Columbia Railway, and a number of regional railways to serve mining and timber resource developments in the north. Although thousands of railway miles have been retired, particularly in the Prairie Provinces, new railroads to northern resources have been constructed, leaving the total track mileage relatively unchanged.

The retirement of track miles is at least partly related to the major decline in the railway share of passenger transportation after World War II as automobile and airplane travel became dominant. In 1977 the Canadian government created VIA Rail, a crown corporation that assumed responsibility for most passenger trains. VIA Rail owns its trains, but it uses the tracks and other facilities of CN and CP. Even though VIA Rail introduced new equipment and improved services, it was not able to stem the tide of declining railway passenger travel. In 1989 its government subsidies were cut and many passenger routes discontinued, including, in 1990, transcontinental service. Most of Canada's railway passenger service is concentrated in the densely populated Windsor-to-Quebec city corridor. In 1967, GO Transit, an agency of the Ontario government, began operating commuter trains in the heavily urbanized area around Toronto. A similar commuter train operation began in the Montreal area in 1984.

Railway passenger traffic

Waterways. A large proportion of goods carried in Canada, in both domestic and international trade, uses water facilities for some part of its journey. The inland shipping routes are dominated by the St. Lawrence–Great Lakes waterway, which provides navigation for vessels of 26-foot draft to the head of Lake Superior, 2,342 miles from salt water. It includes the major canals of Canada. There are seven locks between Montreal and Lake Ontario; the Welland Ship Canal bypasses the Niagara River and Niagara Falls between Lake Ontario and Lake Erie with eight locks; and the Sault Ste. Marie Canal and lock link Lake Huron and Lake Superior. These 16 locks overcome a drop of some 582 feet from the head of the lakes to Montreal. The St. Lawrence Seaway accommodates all but the largest oceangoing vessels, making the upper St. Lawrence and Great Lakes area open to 80 percent of the world's maritime fleet. The main commodities shipped are grain from Thunder Bay on Lake Superior to St. Lawrence ports and iron ore to steel mills in both Canada and the United States.

On the west coast, large volumes of forest products are moved by tug and barge operations. On both the west and east coasts there is an extensive network of ferry services. Shipping is crucial to the development of the Canadian Arctic; it provides a means of transporting mineral resources to markets and bringing supplies to remote communities. In addition to ocean shipping to the Arctic, there is a barge service that transports supplies along the Mackenzie River all the way to its mouth.

By international standards, the Canadian merchant fleet is quite small. Most Canadian-registered merchant vessels operate on domestic routes; only a few Canadian-flag ships operate deep-sea routes. The Canadian Coast Guard ensures that all ships plying Canadian waters, including the Arctic waterways, meet the requirements of the Canadian Shipping Act and follow pollution-prevention procedures.

Airways. Vast distances, rugged terrain, and extreme variations in climate have shaped the development of civil aviation in Canada, and air transport is of tremendous importance. Two major airlines, Air Canada and Canadian Airlines, form the nucleus of Canada's domestic freight and passenger air service. Five major regional domestic air carriers operate other scheduled commercial services. More than 700 small airlines operate nonscheduled services, many of them to parts of Canada that are inaccessible by other means of transportation. Both Air Canada and Canadian Airlines operate international routes.

Major Canadian airlines

The Lester B. Pearson International Airport at Toronto is by far the busiest in the country, handling about 15 million passengers a year in the late 1980s. Montreal has two major airports, the newer one built in 1975 at Mirabel, some 20 miles north of the city.

Pipelines. Pipelines are a major element in Canada's vast transportation network. Growth has been so rapid since 1950, when pipelines were a negligible factor in intercity freight traffic, that oil and gas pipelines now account for about one-fourth of intercity freight ton-miles. Some of the world's longest oil and gas pipelines have been built to link the oil fields and gas fields of Alberta and Saskatchewan to major cities as far east as Montreal, and two major pipelines several hundred miles in length cross the Rocky Mountains and supply the lower mainland of British Columbia and the Pacific Northwest of the United States. The network is to be extended to facilitate transmission of oil and gas south from the Arctic.

Administration and social conditions

Canada is an independent federal parliamentary state, the head of which is the reigning monarch of the United Kingdom. The state was established by the British North America Act of 1867, which united the colonies of Nova Scotia, New Brunswick, and Canada into the provinces of Nova Scotia, New Brunswick, Quebec, and Ontario. Rupert's Land and the Northwest Territories were acquired from the Hudson's Bay Company in 1869, and from them Manitoba was created and admitted to the Confederation as a province in 1870; its extent was enlarged by adding further areas from the territories in 1881 and 1912. The colonies of British Columbia and Prince Edward Island were admitted as provinces in 1871 and 1873, respectively. In 1905 Saskatchewan and Alberta were created from what remained of the Northwest Territories and admitted to the Confederation as provinces. In 1912 the provinces of Quebec and Ontario were enlarged by adding areas from the Northwest Territories. In 1949 Newfoundland and its dependency, Labrador, joined the Confederation following a popular referendum. The Yukon Territory in 1898 was separated from the Northwest Territories. Thus, Canada now consists of 10 provinces and two territories, which vary greatly in size.

Development of the provinces

An act legislated by the British Parliament and proclaimed by Queen Elizabeth II on April 17, 1982, formally ended all vestiges of British control and made Canada responsible for all changes to its own constitution.

GOVERNMENT

The Constitution Act is not an exhaustive statement of the laws and rules by which Canada is governed. The constitution of Canada in its broadest sense includes other statutes of the United Kingdom; statutes of the Parliament of Canada relating to such matters as the succession to the throne, the demise of the crown, the governor-general, the Senate, the House of Commons, electoral districts, elections, and royal style and titles; and statutes of provincial legislatures relating to provincial legislative assemblies. Many of the rules and procedures of Parliament are not laid down in the Constitution Act but are established by convention and precedent, often that of the United Kingdom.

The preservation of both the English and French languages is safeguarded by the provision that either language may be used in all institutions (including the courts) of the Parliament and government of Canada and in all institutions of the National Assembly of Quebec, the legislature of New Brunswick, and their governments. The act also guarantees to Quebec the right to a Roman Catholic school system under Roman Catholic control and to exclusive jurisdiction over property and civil rights and safeguards preservation of the French system of civil law in the province. When the constitution was patriated to Canada in 1982, it was amended to include a Charter of Rights and Freedoms, which provides greater detail than the original charter. Further amendments of the constitution require agreement of the federal government and seven provinces that together represent 50 percent of the country's people. All the provinces approved the constitution except Quebec. The province of Quebec refused to approve the constitution because it infringed on its policy related to the restriction of the use of the English language, did not give Quebec a veto on future constitutional changes, and did not officially recognize Quebec as a distinct society.

Quebec's anti-Constitution Act stand

Federal legislative and executive authority. Federal legislative authority is vested in the Parliament of Canada, which consists of the sovereign of Canada, the Senate, and the House of Commons. The Senate has 104 members, who are appointed on a provincial basis and may hold office until they are 75 years of age. The 295 members of the House of Commons are elected by the people of Canada for maximum terms of five years. Each electoral district elects one member of Parliament to the House of Commons on a universal suffrage basis.

The Constitution Act divides legislative and executive authority between Canada and the provinces. The Parliament of Canada is assigned authority over control of the armed forces, the regulation of trade and commerce, banking, credit, currency and bankruptcy, criminal law, postal services, fisheries, patents and copyrights, the census and national statistics, the raising of money by taxation, and, in the field of communication, such matters as navigation and shipping, railways, canals, and telegraphs. In addition, the federal government is endowed with a residual authority in matters beyond those specifically assigned to the provincial legislatures and including the power to make laws for the peace, order, and good government of Canada. Both the House of Commons and the Senate must pass all legislative bills before they can receive royal assent and become law. Both bodies may originate legislation, but only the House of Commons may introduce bills for the expenditure of public funds or the imposition of any tax.

The personal representative of the crown in Canada, the governor-general, is appointed by the reigning monarch of the Commonwealth upon the advice of the Canadian government. The governor-general exercises the executive authority of the monarch in relation to the federal government. On the recommendation of responsible advisers, the governor-general summons, prorogues, and dissolves Parliament, assents to bills, and exercises other executive functions.

The leader of the party winning the most seats in a general election is called upon by the governor-general to form a government. He becomes the prime minister and generally chooses party colleagues from among the elected members to form the Cabinet. Almost all Cabinet ministers are also heads of executive departments of the government, for the work of which they are responsible to the House of Commons. The Cabinet is responsible for determining all important policies of government and securing the passage of such legislation, financial measures, and administrative provisions as their supporters may approve.

The Cabinet

The ministers of the crown, as members of the Cabinet are called, are chosen generally to represent all regions of the country and its principal cultural, religious, and social interests. Although they exercise executive power, they are collectively responsible to the House of Commons and can remain in office only so long as they retain the confidence of the House. The choice of the Canadian electorate not only determines who shall govern Canada but also, by deciding which party receives the second largest number of seats in the House, designates which of the major parties becomes the official opposition. The function of the opposition is to offer intelligent and constructive criticism of the existing government.

Canada has three major national political parties: Progressive Conservative, Liberal, and New Democratic. In political philosophy, the Progressive Conservatives tend to be slightly to the right, the Liberals in the centre, and the New Democrats on the left. Only the Progressive Conservatives and Liberals have formed national governments. From the 1930s to the 1980s, both the Progressive Conservatives and the Liberals became somewhat more liberal regarding social and health welfare policies and government intervention in the economy. In this way they continued to take votes from the more socialist-oriented New Democratic Party. Under the leadership of Brian Mulroney, who became prime minister in 1984, the Progressive Conservative government underwent a distinctly conservative shift, evidenced by the selling of crown corporations, deregulation of some industries, and the granting of tax advantages to corporations and the wealthy.

Provincial and territorial government. Similar political institutions and constitutional usages operate in the government of the 10 provinces as in that of the nation as a whole. In each province the sovereign is represented by a lieutenant governor appointed by the governor-general in council, usually for a term of five years. The powers of the lieutenant governor in the provincial sphere are essentially the same as those of the governor-general in the federal sphere.

The election of a legislative assembly and the conventions of Cabinet government operate as in the federal House of Commons. The legislature of each province is unicameral and its assembly is elected for five years. The provinces have powers embracing mainly matters of local or private concern such as property and civil rights, education, civil law, provincial company charters, municipal government, hospitals, licenses, management and sale of public lands, and direct taxation within the province for provincial purposes.

Provincial legislatures

The vast and sparsely populated regions of northern Canada lying outside the 10 provinces—Yukon Territory and the Northwest Territories—are administered by the federal government, but they have both elected representation in the House of Commons and a measure of local self-government.

Federal-provincial relations. A major part of Canada's constitutional development has occurred gradually through judicial interpretation and constitutional convention and through executive and administrative coordination at the federal and provincial levels of government. Through such devices, the national and provincial legislatures have been able to retain their separate jurisdictions over different aspects of the same matters. Regular meetings of first ministers are held to discuss federal-provincial jurisdictional issues. Only in such an approach can Canada achieve an accommodation that gives fair assurances to the aspirations of the provinces without disrupting the integrated national structure of the Canadian Confederation.

Of the provinces, Quebec has the most pronounced political aspirations. Only there are citizens of French origin in the majority. Their survival in Quebec as a distinctive and majority group is not difficult to explain in light of the historical constitutional guarantees. They have therefore developed a culture of their own that differs in many respects from that of the rest of Canada—and, indeed, from the rest of North America. The aims of politics for them have been not only material prosperity and liberty in the Anglo-Saxon sense but also ethnic survival. As a consequence they have tended to act as a unit in national matters and support whichever party is paying attention to their claims. The Quebec idea of a distinct society has led to strong movements in favour of political secession of the province from the Confederation. In 1976 Quebec elected a majority government of the Parti Québécois, whose major policy platform was "sovereignty association," a form of separation from Canada but with close economic ties. In 1980 Quebec voted the concept down.

At the national level efforts have been made to create a dual culture in Canada rather than simply preserve two cultures. Thus, the Official Languages Act of 1969 declares that the English and French languages "enjoy equality of status and equal rights and privileges as to their use in all the institutions of the Parliament and Government of Canada." Pursuant to this act, a commissioner of official languages was appointed to ensure compliance with the intent of the act.

Variations in municipal government

Local government. Because government at the municipal level falls under the jurisdiction of the provinces, there are 10 distinct systems of municipal government in Canada, as well as many variations within each system. The variations are attributable to differences in historical development and in area and population density. Thus, the legislature of each province has divided its territory into geographic areas known generally as municipalities and, more particularly, as counties, cities, towns, villages, townships, rural municipalities, or municipal districts. The county system as understood in Britain or the United States exists only in southern Ontario and southern Quebec. County councils are composed of representatives from rural townships, towns, and villages and provide a second level of services for the whole county. This two-tiered municipal government was first extended to urban areas when Metropolitan Toronto was established in 1953. A number of other highly urbanized areas in Ontario have since adopted a metropolitan or regional form of government to deal with common areawide problems. The more than 4,600 incorporated municipalities and local government districts in Canada have various powers and responsibilities suited to their classification. A municipality is governed by an elected council. The responsibilities of the municipalities are generally those most closely associated with the citizens' everyday life, well-being, and protection. In addition to local municipal government, there are numerous local boards and commissions, some elected and others appointed, to administer education, utilities, libraries, and other local services. The sparsely populated areas of the provinces are usually administered as territories by the provincial governments.

Justice. Canadian courts of law are independent bodies. Each province has its police, division, county, and superior courts, with right of appeal being available throughout provincial courts and to the federal Supreme Court of Canada. At the federal level the Federal Court has civil and criminal jurisdictions with appeal and trial divisions. All judges, except police magistrates and judges of the courts of probate in Nova Scotia and New Brunswick, are appointed by the governor-general in council, and their salaries, allowances, and pensions are fixed and paid by the Parliament of Canada. They cease to hold office at age 75. Criminal law legislation and procedure in criminal matters is under the jurisdiction of the Parliament of Canada. The provinces administer justice within their boundaries, including the organization of civil and criminal codes and the establishment of civil procedure.

Royal Canadian Mounted Police

Police and the military. The police forces of Canada are organized in three groups: the federal force, called the Royal Canadian Mounted Police (RCMP), the provincial police, and municipal police. The RCMP—or Mounties, one of Canada's best-known organizations—was established in 1873 for service in the Northwest Territories of that time. It is still the sole police force in the Yukon and Northwest territories, but it also has complete jurisdiction of the enforcement of federal statutes throughout Canada, which includes the control of narcotics. The maintenance of peace, order, and public safety and the prevention and investigation of criminal offenses and of violation of provincial laws are provincial responsibilities. Ontario and Quebec have their own provincial police forces, but all other provinces engage the services of the RCMP to perform these functions. Provincial legislation makes it mandatory for cities and towns and for villages and townships with sufficient population density and real property to furnish adequate policing for the maintenance of law and order in their communities. Most large municipalities maintain their own forces, but others engage the provincial police or the RCMP, under contract, to attend to police matters. In 1984 the Canadian Security Intelligence Service (CSIS) was created to replace the security service previously provided by the RCMP. The purpose of the CSIS is to conduct security investigations within Canada related to subversion, terrorism, and foreign espionage.

Matters relating to national defense, including the armed forces, are the responsibility of the minister of national defense. Canada's armed forces constitute a considerably lesser percentage of the Canadian labour force than do the armed forces of its allies in the North Atlantic Treaty Organization (NATO), and its defense spending is lower per capita than that of most of its allies. Military strength has never been central to Canada's national security efforts. Instead, it has worked through the United Nations at keeping peace and has formed strong alliances with the United States and NATO.

EDUCATION

Under the Constitution Act the organization and administration of public education are provincial responsibilities. The federal government is directly concerned only with the provision of education in the Yukon and Northwest territories, in Indian schools throughout Canada, for inmates of federal penitentiaries, for the families of members of the Canadian forces on military stations, and for the operation of three military colleges. In addition, it finances vocational training of adults and provides financial support to the provinces for the operating costs of postsecondary education.

Because each of the 10 provinces has the authority and responsibility for organizing its education system as it sees fit, policies and practices vary from province to province, but each has a department of education headed by a minister who is a member of the provincial Cabinet. In addition, Ontario has a Department of Colleges and Universities with a minister of its own. Most Canadian children have one year of kindergarten before they enter an eight-grade elementary school at age 6 or 7. At about 14 years of age, nearly 90 percent of those who entered grade one enter a regular four-year secondary school.

Higher education

Traditionally, a higher education was chiefly the preserve of universities. Now, although universities still account for about 60 percent of full-time students, postsecondary education is available in institutions without degree-granting status—regional colleges in British Columbia, institutes of technology in Alberta, institutes of applied arts and sciences in Saskatchewan, colleges of applied arts and technology in Ontario, and *collèges d'enseignement général et professionel* (community colleges) in Quebec.

Canada has more than 65 degree-granting institutions. They range from institutions with a single faculty and enrollments of a few hundred to institutions with many faculties and research institutes and more than 40,000 students, such as the Université du Québec (a multicampus university), the University of Toronto, and the University of British Columbia.

The oldest university in Canada, Laval, in Quebec, was founded during the French regime. Universities in English-speaking Canada were established after the American Revolution. University of King's College in Nova Scotia and what is now the University of New Brunswick were patterned after King's College (now Columbia University) in pre-Revolutionary New York City. Most other universities in pioneer days were begun by churches, but almost all have since become secular and almost entirely financially dependent on the provincial governments. In the 1960s Ontario established a number of new postsecondary institutions. One of the new universities, the University of Waterloo, has a cooperative program (alternating academic and work terms) and has gained an international reputation in mathematics, computer science, and remote sensing. There are no truly private universities in Canada. A somewhat unusual characteristic has been the system of "affiliated colleges" linked to a "parent" degree-granting institution even though separated from it physically. English is the language of instruction at most places, French in several, and a few are bilingual.

HEALTH AND WELFARE

Canada has a diversity of joint federal-provincial and local programs to meet the health and social needs of its people: health insurance; general public-health services, primarily of a preventive nature; services for specific diseases or disabilities, combining prevention, treatment, and rehabilitation related to general medical and hospital care; and welfare services. The health professions and voluntary agencies and institutions supply services, initiate new activities, and stimulate better standards. The federal government provides consultative, coordinating, and financial assistance. The federal Medical Care Act (1968) made provision for federal contributions to medical-care plans of participating provinces provided that such plans are comprehensive with regard to medical services, have universal coverage, are administered by a public authority, and allow portability of benefits between provinces. The federal Hospital Insurance and Diagnostic Services Act (1957) allowed for federal contributions to the costs of specified hospital services and now covers 99.5 percent of the insurable population of Canada. The federal government also has responsibilities for the administration of food and drug legislation (including narcotics control), quarantine, immigration and sick-mariners services, and the health and welfare of Indians, Inuit, and past and present members of the Canadian armed forces.

National health care provisions

There are a number of social security and social assistance programs. The universal Family Allowance Act, which now provides monthly payments to help support all dependent children up to the age of 18 years, has been a unique feature of the Canadian social security system since its inception in 1945. The Canada Pension Plan provides retirement, disability, and survivors' benefits. The Old Age Security Act provides a monthly pension to all persons 65 years of age or over, while the guaranteed-income supplement provides additional income for old-age pensioners. Financial aid is available under provincial or municipal auspices to persons in need and their dependents. The unemployment insurance system is financed by premiums paid by employers and employees, along with federal government contributions. To receive benefits, applicants must have been previously employed for a stated period and claim that they are actively seeking employment.

Cultural life

The culture of the English-speaking Canadian people is a blend of British and American influences; that of the French-speaking people blends French and American influences. In general, the way of life, family structure, cuisine, and dress are closer to those of the United States than to those of Britain or France. Canada has become a cultural mosaic in which immigrant groups have been able to retain much of their ethnic culture.

The British and French heritage, plus the influence of the United States, has led to broad diversity of artistic endeavour in Canada, a diversity further encouraged by groups of other national origins and a widely dispersed population. The major metropolitan centres have provided several focal points for artistic activity stimulated by a rapidly urbanizing society.

Since 1950 economic growth has provided all Canadians with greater means for the practice and enjoyment of the arts, and the influx of immigrants from many countries has increased both the audience and the pool of available talent. Most provincial governments provide some form of financial assistance for the arts and for cultural organizations within their borders; and some have advisory and funding councils for the arts. At the national level the Canada Council was established in 1957 "to foster and promote the study and enjoyment of, and the production of works in the arts, humanities and social sciences." It is funded by an endowment, by an annual grant from the federal government, and by private donations. The annual Governor General's Literary Awards, the preeminent literary prizes in Canada, have served to reward Canadian writers as well as to publicize Canadian literature through ceremonies held in various centres across the country.

LITERATURE

French literary heritage

The first truly Canadian literary works were written in French by explorers, missionaries, and settlers, and many of them became the inspiration of subsequent writings. Some were notable literature, such as Marc Lescarbot's *Histoire de la Nouvelle France* (1609). The first major contribution in English was made by Thomas Haliburton of Nova Scotia, with his *The Clockmaker; or, The Sayings and Doings of Samuel Slick of Slickville* (1836). The years following were also marked by some works that have become classics—William Kirby's *Golden Dog* (1877), Robert W. Service's *Trail of '98,* the humorous works of Stephen Leacock, and the long series of Jalna novels by Mazo de la Roche. Natural and vigorous French-Canadian writing was held back by a sense of isolation, strict Roman Catholicism, a small and relatively unsophisticated audience, and a tendency to favour academic and formal literary styles and themes. The 1940s witnessed the emergence of Canadian writers of the first rank. Hugh MacLennan established himself as a writer of international importance with *Barometer Rising* (1941) and *Two Solitudes* (1945), Thomas Raddall with *His Majesty's Yankees* (1942), and W.O. Mitchell with *Who Has Seen the Wind?* (1947). Gabrielle Roy's novel *Bonheur d'Occasion* (1945) was an immediate success, and *Le Survenant* (1945), by Germaine Guèvremont, placed its author in the forefront of French-language novelists, in both Montreal and Paris. Still later came the novels of Robertson Davies and the satires of Mordecai Richler. The French-Canadian novel, long handicapped by sentimentality and stilted prose, went beyond even hard realism into the psychological realm with Marie-Claire Blais's *La Bell Bête* (1959). In the 1960s and '70s other important authors, among them Margaret Laurence, Alice Munro, and Margaret Atwood, emerged to international prominence and made significant contributions on into the 1980s. In the 1980s Davies wrote a successful trilogy of novels, and Richler produced his most ambitious work, *Solomon Gursky Was Here* (1989). Although the growth of competent novel writing was the main feature of Canada's literary scene after World War I, the period saw marked changes in the work of Canadian poets. John McCrae's *In Flanders Fields* (1919) was the only important Canadian verse related to the 1914–18 conflict, but since then E.J. Pratt, Earle Birney, Irving Layton, Anne Hébert, James Reaney, Al Purdy, and Ralph Gustafson, among others, have attracted widespread attention.

VISUAL ARTS

Painting

From the beginning, artists in Canada have been greatly influenced by the art of their European roots. At the same time, Canadian painting has been successful in expressing a distinctive Canadian national identity. This has been particularly true of a long succession of artists who have portrayed Canadian locales and landscapes. In the 19th century Paul Kane, an immigrant from Ireland, traveled across Canada and painted numerous canvases depicting Canadian landscapes and the lives of Indians, fur traders, and missionaries, all rendered in a contemporary European genre style. During the same period, Cornelius Krieghoff, of German descent, painted more than 2,000 canvases of anecdotal scenes in Quebec. His paintings brought new dimensions to the Canadian scene and a colourful romanticism unsurpassed by contemporary artists. The first distinctively Canadian painting, free of Old World influences, is generally attributed to Homer Watson, an Ontario landscape painter of the late 19th and early 20th centuries. In the 1920s an assemblage of painters called the Group of Seven sought to develop a national form of painting, taking their inspiration from the Canadian landscape and the work in particular of the Canadian landscape painter Tom Thomson. A.Y. Jackson was among the best known of this group. Emily Carr, a contemporary of the Group of Seven, depicted Indians of the west coast and was heavily influenced by their art.

After the 1930s Canadian painters generally moved away from the landscape theme. In Quebec, where art tended to be more theoretical than in the other provinces, painting evolved through a number of movements. In reaction to automatism, Guido Molinari and Claude Tonsignant freed painting from its surrealism and directed it toward the concern of structure and colour. Similar trends occurred in Ontario, where a group called Painters Eleven, led by Harold Town and Jack Bush, moved from abstract expressionism to formalism.

Contemporary art in Canada is marked by a questioning of the nature of art, accompanied by experimentation and innovation. The resulting work ranges from the intensely personal to the public airing of concern about social and environmental issues.

Inuit sculpture and carvings

Although painting has received most of the attention in the art world of Canada, sculpture and the handicrafts have been in evidence since earliest times and only now are receiving adequate public attention. Prominent in sculpture exhibitions are stone carvings of the Inuit people, while among the most interesting of museum-guarded sculptures are the totem-pole carvings of the Indians of the west coast. In the European-derived cultures of Canada, the development of sculpture lagged behind that of painting. For centuries, sculpture meant carved figures made of wood, stone, or bronze. It was not until the 1960s that sculptors challenged the traditional notions of form, content, and technique. Artists such as Les Levine and Michael Snow were also painters, but their three-dimensional work established their reputations. By the 1970s and '80s there was established a new tradition of constructed sculpture in which abstract shapes were created from a variety of materials, including steel, aluminum, and plastic. Canadian sculptors were beginning to collaborate with architects in the design of public buildings. Architecture in Canada tends to be international in style, but some architects, such as Arthur Erickson and John Andrews, an Australian, have designed buildings of elegance that reflect truly Canadian influences.

GALLERIES AND MUSEUMS

Along with developments in the visual arts came the establishment of art collections and art galleries. The National Gallery of Canada in Ottawa, dating from 1880, includes not only the most extensive and important collection of Canadian art in existence but also collections built up along international lines to give an indication of the origins from which Canadian tradition is developing. It also circulates exhibitions to several hundred centres in the country each year. In addition, major cities have public art galleries, and many arts councils and university galleries house important collections.

Many museums in Canada keep Canadian historic artifacts on public display. There are several national museums on specific themes located in Ottawa, and almost every city and town has a local museum. The Royal Ontario Museum, Canada's largest museum, is visited by more than one million people annually. There are also many historic parks and monuments in Canada, the most ambitious being the 13,000-acre site around the reconstructed Fortress of Louisburg in Nova Scotia.

THE PERFORMING ARTS

There was a virtual explosion of musical activity in Canada in the second half of the 20th century. Choral music societies sprang up across the country. Opera grew, with Toronto, Montreal, and Vancouver having regular opera seasons and the Toronto-based Canadian Opera Company making extensive tours, often to remote parts of the country. Many cities have their own symphony orchestras, particularly those with Canadian Broadcasting Corporation (CBC) broadcast centres where musicians can be sustained by radio and television assignments. The symphony orchestras of Toronto and Montreal are internationally recognized. International acclaim has also been won by musical groups such as the Orford String Quartet, the Festival Singers of Canada, and the Canadian Brass.

Individual Canadian performers who have received world acclaim include the singers Maureen Forrester and Lois Marshall, the pianists Glenn Gould and Austrian-born Anton Kuerti, the guitarist Liona Boyd, and jazz musicians Maynard Ferguson and Oscar Peterson. Well-known popular and folk singers and songwriters include Gordon Lightfoot, Anne Murray, Gilles Vigneault, Bruce Cockburn, Leonard Cohen, and k.d. lang.

Dance

Canada can claim three top-ranking professional ballet companies: the Royal Winnipeg Ballet (founded in 1939), the National Ballet of Canada (founded in 1951), and Les Grands Ballets Canadiens (founded in 1957). Les Ballets Jazz de Montréal is a dynamic dance company that combines ballet technique and jazz music. The National Ballet is the largest and most widely traveled, touring across Canada, the United States, and Europe.

Canada's professional theatre evolved out of the amateur Little Theatre movement, which involved Canadian playwrights and performers and which developed a knowledgeable and appreciative audience. Several year-round repertory groups in the largest cities became professional. In 1953 the Stratford Festival was founded in Stratford, Ont., and it became an immediate success, drawing people from across Canada and the United States to see performances of Shakespeare's plays. Another celebrated summer theatre is the Shaw Festival at Niagara-on-the-Lake, Ont. The Blyth Festival in rural southwestern Ontario specializes in Canadian plays on rural issues. In Quebec during the 1960s and '70s, drama expressed the Québécois society's social and political aspirations. By the 1980s French Canadian theatre was again concerning itself with broader universal issues.

Although most Canadian amateur and professional musical theatre groups prefer to present Broadway musicals, Canadians continue to compose musicals on Canadian topics. The Charlottetown Festival produces Canadian shows exclusively. Charlottetown's most successful show, *Anne of Green Gables,* has been produced both in London and on Broadway.

The enthusiasm engendered by the success of theatre ventures set off a new Canadian determination to have professional theatre on a regular and nationwide scale. Spectacular new theatres were built across the country after 1958, among them Confederation Centre in Charlottetown (Prince Edward Island); Place des Arts (Montreal), National Arts Centre (Ottawa); O'Keefe Centre, St. Lawrence Centre for the Performing Arts, and Pantages Theatre (Toronto); the Centennial buildings of Winnipeg and Saskatoon; the Jubilee theatres of Edmonton and Calgary; and the Queen Elizabeth Theatre in Vancouver.

FILMMAKING

The National Film Board

The impact of television on the film production industry can be readily assessed by the fact that about three-quarters of the films produced as features, advertising, trailers, newsclips, and newsreel stories by Canada's private and governmental filmmaking agencies are for television. The National Film Board was established by the federal government in 1939 to produce films, filmstrips, and still photographs that reflect the life and thought of Canada and to distribute them both domestically and abroad. It has earned international acclaim for the imaginativeness as well as the artistic and technical excellence of its work, winning awards from film festivals around the world. In 1967 the federal government established the Canadian Film Development Corporation to foster and promote a feature-film industry through investment in productions, loans to producers, and grants to filmmakers.

RECREATION

Canadians, enjoying ever-increasing amounts of leisure time, are able to participate in a wide array of sports and other recreational activities.

Sports. Several of the sports played in Canada are derived from those of the indigenous peoples or the early settlers. Lacrosse, adopted as Canada's national game at the time of Confederation, was played by Indians in all parts of the country and adopted by later immigrants. By 1867 definite rules had been established, and the game had become organized. Ice hockey is also Canadian in tradition and leadership. It remains one of Canada's most popular winter sports. The original teams of the National Hockey League, established in 1917, were all Canadian. "Hockey Night in Canada" broadcasts, originating with Foster Hewitt on radio in 1931 and still carried by CBC television, is popular with millions of Canadians. Other team sports have been more strongly influenced by the recreational interests of the United States. The Canadian Football League (CFL) plays a football game only slightly different from that of the United States. The annual Grey Cup game that decides the CFL championship has become a national event and has the largest television audience of any Canadian sporting event. Two Canadian teams, the Montreal Expos and the Toronto Blue Jays, compete in United States professional baseball leagues. Winter sports widely enjoyed by Canadians, as both participants and spectators, include curling, ice skating, and downhill and cross-country skiing. From spring through fall, recreational activities include fishing, hunting, hiking, golf, and water sports.

National parks and historic sites. Canada's vast national park system, which extends to both coasts, began around the hot springs centred in Banff National Park in Alberta, established in 1885. Tourists and Canadian vacationers are attracted to the park sites throughout the year to view the striking natural scenery and to partake of the numerous recreational activities offered. Programs include guided tours, lectures, and films.

In 1919 the federal government established the Historic Sites and Monuments Board to advise on matters of national historic importance, with particular reference to commemoration or preservation. Most of the national historic parks are military or fur-trading forts, historic buildings, or reconstructions of historic buildings; most have museums. They range from the 13,000-acre site around the reconstructed Fortress of Louisburg in Nova Scotia to the boyhood home of former prime minister W.L. Mackenzie King and a reconstructed theatre of the gold-rush days in Dawson, Yukon Territory. Two are preserved Yukon riverboats. Each province has a similar preservation policy. In the west many of the marked sites, such as Head-Smashed-In Buffalo Jump in Alberta, recall the Indian period. Several provinces have restored or reconstructed pioneer communities. British Columbia's first provincial historic park was the restored gold-mining town of Barkerville. In Ontario, Fort Sainte-Marie, the headquarters of the Jesuit mission to the Huron Indians in the 17th century, has been restored.

Provincial preserves

PRESS AND BROADCASTING

Press. In every Canadian city there is a daily newspaper; in the largest cities there is usually more than one. Smaller towns are served by weekly newspapers. The trend in daily newspaper publication is toward chain ownership, but editorial policy is developed at the local level. The Toronto *Globe and Mail,* which calls itself Canada's national newspaper, is distributed six days a week across the entire country. Supporting the newspapers are two news-gathering agencies, the Canadian Press and the United Press International of Canada. Several hundred magazines are published in Canada, but most of the popular magazines that circulate in Canada are from the United States. *Maclean's* is a weekly Canadian newsmagazine.

The market for Canadian books is small by international standards, and thus Canadian publishers have had difficulty meeting foreign competition. Increasingly, Canadian book publishing has been taken over by foreign firms, primarily from the United States.

Broadcasting. In a huge country that includes groups of people of diverse origins separated by vast distances, broadcasting is not only important as entertainment but is also crucial for linking the various regions together to develop a sense of national community. It is for this reason that Canada has developed an elaborate structure and organization for delivering radio and television broadcasts. Canada was the first country in the world to use geostationary satellites for television broadcasting. The publicly owned Canadian Broadcasting Corporation (CBC) provides two national networks for both radio and television, one in English and one in French. The CBC also has a special northern shortwave radio network to provide broadcasts to the more remote Arctic settlements in both official languages as well as in Indian and Inuit-Aleut languages. In television broadcasting there is a second national network, the privately owned CTV, as well as several other private networks serving limited areas, such as Global TV in Ontario. Private radio and television stations usually are affiliated with one of the major networks. As a result, all Canadians have access to radio, and almost all have a choice of two Canadian television channels. Cable and satellite connections have also given Canadians access to U.S. television networks.

Canadian broadcasting regulation

Canadian broadcasting is regulated by the Canadian Radio-Television and Telecommunications Commission (CRTC). It authorizes the establishment of networks and private stations and specifies how much of the broadcast content must be Canadian in origin. The CBC has played an important role in developing Canadian talent in the entertainment world. Its music, drama, and documentary programs are considered to be of high quality.

(N.L.N./R.R.K.)

For statistical data on the land and people of Canada, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

The first inhabitants of the territories that now make up Canada were the Indians and the Eskimo, or Inuit. The former came from northern Asia across the Bering Strait, while the latter were part of the circumpolar dispersion of a race that lived along the edge of the Arctic ice. There were marked differences in cultural development among the Indians, whose population had reached about 200,000 at the beginning of the period of European exploration of the New World in the late 15th century. The west-coast Indians of British Columbia, for instance, tended to group in settlements, while those of the plains and eastern woodlands tended to follow a more migratory pattern. The Inuit were far fewer in number, and, although they lived in various-sized groups, they shared a common culture based largely on hunting. The Inuit group life did not necessarily evolve into a community pattern of life beyond that of the immediate family. For the prehistory of the North American Indians, see AMERICAN INDIANS; for that of the Inuit, see ARCTIC, THE.

Canada to 1867

The first Europeans to reach what is now Canadian soil were Norse explorers about AD 1000, as is recorded in the Icelandic sagas and recently confirmed by the archaeological discovery of a Norse site at L'Anse-aux-Meadows on the northern peninsula of Newfoundland. While the Norse made no lasting settlements, their voyages traditionally begin most general histories of Canada, largely because of their discovery of the northern route to North America.

The voyage of John Cabot

Only five years after the first voyage of Columbus, an Italian navigator, known to the English as John Cabot, sailed from Bristol in 1497 under the patronage of Henry VII of England by the northern route in an attempt to reach Cathay (China). His trip to a landfall generally accepted as having been on the east coast of Newfoundland was made so expeditiously that it is tempting to assume that he had the aid of local knowledge of the route and its winds, perhaps from Bristol fishermen accustomed to sailing in

Iceland's waters. Cabot, however, perished in an unknown manner with a large expedition in the following year.

THE SETTLEMENT OF NEW FRANCE

Jacques Cartier. The discoverer of the great entry to Canada, the St. Lawrence River, was the Frenchman Jacques Cartier. In 1534, in a voyage conducted with great competence, he explored the Gulf of St. Lawrence and claimed its shores for the French crown. In the following year, Cartier ascended the river itself and visited the sites of Stadacona (Quebec) and Hochelaga (Montreal). So favourable were his reports that the French king, anxious to challenge the claims of Spain in the New World, decided to set up a fortified settlement. Internal and European politics delayed the enterprise until 1541, when, under the command of the Lord de Roberval, Cartier returned to Stadacona and founded Charlesbourg-Royal just above Quebec. Cartier had hoped to discover precious gems and minerals, as the Spaniards had done in Mexico and Peru. But the mineral specimens he sent home were worthless, and "false as a Canadian diamond" became a common French expression. Disappointed in his attempt to reach the mythical "Kingdom of Saguenay," the reputed source of precious metals, Cartier returned to France after a severe winter, deserting Roberval, who had arrived in Newfoundland with reinforcements. Roberval also failed, and during the century only two subsequent attempts were made at exploiting the French claim to the lands of the St. Lawrence. The French claim remained; it had only to be made good by actual occupation.

Although French and English fishermen began to sail to the cold Atlantic waters off Newfoundland shortly after Cabot's discoveries, it was not until the early 1600s that France attempted to establish a permanent settlement in North America. That enterprise was based on the granting of fur-trade monopolies, because furs were in great demand in Europe and the Indians seemed eager to trade them for European manufactured goods.

Samuel de Champlain. In 1604 the French navigator Samuel de Champlain, under the Lord de Monts, who had received a grant of the monopoly, led a group of settlers from the St. Lawrence region to Acadia. He chose as a site Dochet Island in the St. Croix River, on the present boundary between the United States and Canada. But the island proved unsuitable, and in 1605 the colony was moved across the Bay of Fundy to Port Royal (now Annapolis, N.S.). The colony was to be a trading post and a centre of settlement, but the rugged, forested inlets of the Nova Scotian peninsula, the heavy forests of the St. John River, and the many bays and beaches of Cape Breton and Prince Edward islands made it impossible to enforce the monopoly of the fur trade against enterprising interlopers.

Foundation of Quebec

In 1608, therefore, de Monts and Champlain left Acadia and returned to the St. Lawrence. At "the place where the river narrowed" (Quebec), they built a "habitation" (*i.e.,* a fur-trading fort, or factory) to control the great river and to be the mart of its fur trade. Already in 1603 Champlain had noted that the Iroquois (or Huronian) Indians, whom Jacques Cartier had found there, had withdrawn from the St. Lawrence under pressure from the Algonquin Indians of the north country and Acadia. The French then became the allies of the Algonquin in the rivalry that began for control of the inland fur trade. In 1609, in accordance with this alliance, Champlain and three companions joined an Algonquin war party in a raid against the Mohawk, the easternmost tribe of the Iroquois League. The party ascended the Richelieu River to Lake Champlain. In an encounter with a Mohawk band, the firearms of Champlain and his men killed some Iroquois and panicked the remainder. This skirmish signaled the initial commitment of New France to the side of the Algonquin and the Hurons (who were Iroquoian but hostile to the confederacy) in the century-long struggle for control of the output of furs from as far away as the western Great Lakes. That commitment deepened in succeeding years. The conflict between the Iroquois and Hurons was based on trade rivalries that had existed since before European settlement. While French support went to the Huron, the Dutch, followed by the English, sided with the Iroquois.

The company of de Monts and his frequent successors, for whom Champlain remained the lieutenant in New France, had the obligation to bring out settlers, as well as the exclusive right, seldom enforced, to trade in furs. Their efforts at settlement were even less successful, partly because settlement was not easy in a country of heavy forests and severe winters and partly because the fur trade had little need of settlers beyond its own employees. The company, moreover, had scant funds to bring out and establish colonists on the land. Champlain did the best he could, and he also encouraged missionaries (first the Recollects [Franciscans], then the Jesuits) to come to Quebec to convert the Indians. His greatest interest, however, lay in exploration. Already in Acadia, in 1606 and 1607, he had surveyed the coast southward and westward to Stage Harbor, only to be rebuffed by hostile Indians.

In 1613 Champlain set out from Quebec to explore the upper St. Lawrence Basin. He passed Montreal Island, not settled since Cartier's time but used by traders who bypassed Quebec. In order to avoid the heavy rapids of the St. Lawrence, he ascended its great tributary, the Ottawa, only to be turned back at Allumette Island by the Algonquin middlemen who were trading for the furs of the Hurons and other people farther inland and who wished to retain that trade. At Allumette, Champlain, however, heard of the "inland sea" (Hudson Bay), the existence of which he had divined before he could have heard of Henry Hudson's discovery of it in 1610. Undaunted, he ascended the Ottawa again in 1615, traversed the Mattawa River, Lake Nipissing, and the French River to Georgian Bay, and turned south to "Huronia" (the land of the Hurons). He wintered with the Indians and went with a Huron war party to raid an Onondaga village south of the St. Lawrence. He was slightly wounded and the party was repulsed, but Champlain had once more confirmed the alliance of the French with the northern tribes and the Hurons against the Iroquois and, by the opening of the Ottawa route, had secured the midcontinent for the French fur trade.

Alliance against the Iroquois

The discovery of this inland, central region was perhaps Champlain's main achievement. He had no success, however, from 1616 to 1627 in maintaining the fur trade. The fault was not entirely his, for the enterprise in itself was very difficult. The coupling of trade and settlement was somewhat contradictory, and it was impossible to finance both out of annual profits, especially as the French government failed to uphold the monopoly.

The Company of New France. The French government was to supply more active support after the remarkable revival of royal power carried out by Cardinal de Richelieu in the 1620s. Richelieu sought to make French colonial policy comparable with that of England and the Netherlands, joint victors with France in the long struggle in Europe against Spain. These countries had found a means of both raising capital and enforcing trading rights through the medium of the joint stock company. Richelieu used his power to create such a company to exploit the resources and settle the lands of New France. This was the Company of New France, commonly called the "Hundred Associates" from the number of its shareholders. It was given broad powers and wide responsibility: the monopoly of trade with all New France, Acadia as well as Canada; powers of government; the obligation to take out 400 settlers a year; and the task of keeping New France in the Roman Catholic faith.

The company was chartered and its capital raised in 1627. The next year, however, war broke out with the English, who supported the French Protestants, or Huguenots, in their struggle against Richelieu. The war was mismanaged and inconclusive, but it gave a pretext for the Kirke brothers, English adventurers who had connections in France with Huguenot competitors of the Hundred Associates, to blockade the St. Lawrence in 1628 and to capture Quebec in 1629. For three years the fur trade was to be in the hands of the Kirkes and their French associates the brothers de Caën. It was a stunning blow to the new company and to Champlain, who was taken prisoner to England. At the same time, Acadia, already raided from Virginia in 1613, was claimed by Scotland. An attempt at settlement

The Scots in Nova Scotia

there was made by Sir William Alexander, to whom Nova Scotia (New Scotland) had been granted by the Scottish king James VI (after 1603, James I of England).

It is difficult to estimate the effect of the war on the policy of the Hundred Associates. Canada and Acadia were restored by the Treaty of Saint-Germain-en-Laye in 1632, and the company retook possession in 1633. On the surface all seemed to go smoothly. In 1633 Champlain returned as governor; the government and settlement of Acadia was farmed out to the vigorous Isaac de Razilly; the Society of Jesus assumed sole responsibility for the care of souls in Canada and for the conversion of the Indians. The fur trade was resumed, and the Trois Rivières settlement was founded in 1634 to control the Saint-Maurice River. Settlement began, but the company seemed unable to recoup the losses caused by the capture of Quebec and by the disruption of five years' trade and seemed unable to make profits that would both pay dividends and provide for the costs of settlement. The company remained the proprietor of New France until 1663, providing a succession of governors and other officials, but it never succeeded in meeting its obligations to colonize. Weary of its profitless task, the company leased the fur trade to private companies and then, in 1645, to a group of Canadian residents known as the Community of Habitants.

The character of French settlement. The fur trade was not the sole enterprise of New France. By 1645 settlers in Canada and Acadia were producing provisions for the fur traders and the annual ships. A characteristic mode of landholding, known as the seigneurial system, began to evolve; it was essentially a modified form of feudal French land law and justice. But the great partner and sometime rival of the fur trade was the missionary endeavour of the Jesuits. They had two obligations: (1) to keep New France Catholic by ministering to its people and excluding Huguenots and (2) to convert the Indians. The missionaries made the conversion of the agrarian Hurons their principal concern. Huronia was the hub of the inland fur trade. To make Huronia a Christian community would create a centre of Christianity and confirm the French commercial alliance with the Hurons and their Algonquin clients. French missionaries had already visited Huronia in the mid-1620s, and in 1634 the Jesuits resumed the mission, which thrived, at least outwardly, for 10 years.

Increased Iroquois hostilities

As the French-Huron alliance tightened, Iroquois hostility toward both parties increased. This was a case of traditional tribal trade rivalries being exacerbated by newer trade rivalries involving Europeans. The Iroquois sought to eliminate the Hurons and take complete control of the interior fur trade. They were able to obtain firearms from the Dutch in the Hudson River valley and launched ever more devastating raids on Huronia. The French tightly controlled the firearms trade with their Huron allies, putting the latter at a tremendous disadvantage. In 1648 and 1649 the Iroquois inflicted major defeats on the Hurons, virtually eliminating them as a major factor in the region.

These checks to both the fur trade and the missions, at least in terms of the intentions and hopes of 1627, were the result not only of bad luck and poor management but also of the economic conditions of New France, which depended almost entirely on the fur trade for profit. Settlement was unprofitable to both the company and the colonists. The population of New France, therefore, grew quite slowly, rising from an estimated 200 residents (*habitants*) in 1642 to perhaps 2,500 by 1663; and by no means were all of these farmers. The fur trade, however, was booming, borne up by the fashion of the beaver felt hat in Europe. The traders brought French goods to trade with the flotillas of canoes that carried the furs of the Ottawa and Great Lakes countries, which before 1648 were usually manned by Huron middlemen. This was the sole commercial enterprise of New France.

Royal control. New France, though a proprietary colony, was governed by the company, which appointed governors for Canada and Acadia, and a few dependent officers. The kings of France remained interested in the colony, both because of the vast potential wealth of the area and because the crown might have to resume the powers of government given to the Hundred Associates. Government was, in fact, very much what it would have been if the colony had been directly under the rule of the crown. In 1647 a council was established in New France that included the governor, the chief religious authority, the superior of the Jesuits, and the governor of Montreal. During the brief rule of the Community of Habitants, representatives (syndics) of the people of Quebec, Trois Rivières, and Montreal were consulted on local matters. This, however, was the nearest approach to anything like representative government. Government in New France, as in old, was authoritarian and paternalistic.

Colbert's reorganization

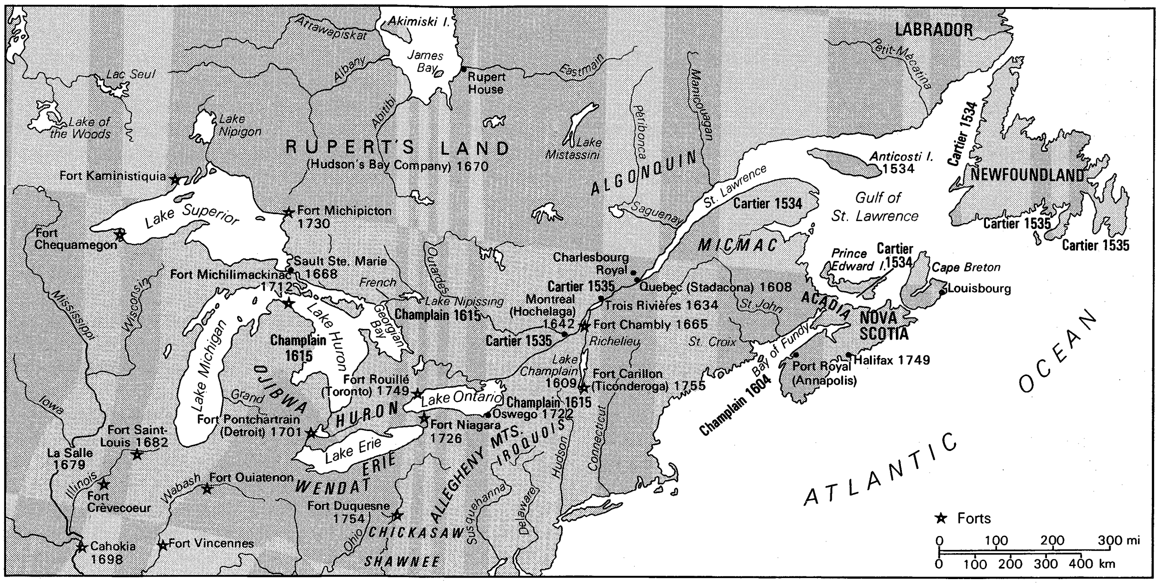
The assumption of power by Louis XIV, however, and the colonial ambitions of his great finance minister, Jean-Baptiste Colbert, led to a recasting of French colonial policy and of the government of New France. Colbert entrusted commercial policy to a new Company of the West Indies. Politically, he made New France a royal province, governed much like a province of France itself. It was to be governed by three persons: a governor, an intendant, and a bishop. The governor was the largely titular head of this triumvirate, although he was responsible for matters of defense and relations with the Indians. He was aided in his decision making by the Superior Council (first called the Sovereign Council), which was to advise him during the long periods when he had no communication with France. The intendant was responsible for internal matters, and the bishop administered mission work and the church. Both the intendant and the bishop were members of the council. Bitter rivalries were not unknown among these officials.

The general effect of Colbert's reorganization, however, was to give New France firm and rational government thereafter, strongly centralized and efficient for the times. The exception was Acadia. Torn by feuds among French rivals, claimed by England, and occupied by New Englanders eager to exploit its fishery, Acadia did not again become an effective part of New France until 1667–70.

The strength of the royal government was in inverse proportion to the weakness of a small and scattered population. Great efforts made by the first intendant, Jean Talon, did indeed bring some thousands of settlers, hundreds of them women, to New France in the 1660s and early 1670s. The population in 1666 reached 3,215; in 1676, about 8,500; thereafter, however, the population grew largely by natural increase, fortunately at a prodigious rate. Most of the population lived in the three towns (Montreal, Quebec, and Trois Rivières) and in seigneuries along the banks of the St. Lawrence between Quebec and Montreal. Scores of the men, however, went inland with the trading canoes, and some of these voyageurs remained inland permanently, marrying Indian women and fathering the métis, or people of mixed French and Indian blood.

The frontier of New France was not a broad front of advance but, rather, a penetration of the wilderness via the rivers in search of furs and strategic position. Alliances with Indians continued to be necessary, and those alliances were constantly challenged by the Iroquois, who controlled the region south of Lakes Ontario and Erie in the 1650s. Thus war with the Iroquois continued, as did the push into the interior, Louis Jolliet penetrating via the Mississippi River as far as the confluence with the Arkansas River.

The growth of Anglo-French rivalry. In the 1660s two voyageurs, Médard Chouart des Groseilliers and Pierre Esprit Radisson, exasperated by the high cost of the long haul back to Quebec and by the heavy tax on fur pelts, fled to New England. From there they were escorted to England, where in 1668 they persuaded a group of London merchants to attempt to gain the fur trade of the mid-continent by way of Hudson Bay. This led to the formation in 1670 of the Hudson's Bay Company, a late proprietary company that was given exclusive trading rights in all the territory draining into Hudson Bay. New France now found itself caught between the Iroquois, supported by the Dutch and English, to the south and the Hudson's Bay Company to the north. The Count de Frontenac, the governor of New France, after his arrival in 1672 made a vigorous push into the continental interior. Frontenac



New France, 16th–18th centuries.

had been directed to concentrate settlement in areas with easy sea access to France, but he defied those instructions in search of profits from furs. For this and other transgressions he was recalled in 1682.

The Treaty of Utrecht

Over the next three decades the French struggled—sometimes with success—to improve their strategic position in America. The British were almost completely expelled from Hudson Bay by 1700, while in the late 1690s Frontenac (who returned as governor in 1689) finally defeated the Iroquois, who sued for peace. Much of this success was lost, however, by the Treaty of Utrecht, which ended Queen Anne's War (1702–13) between the British and French in North America, as well as the War of the Spanish Succession. By that treaty France lost its claim to Hudson Bay, its hold on Acadia, and its position in Newfoundland. After Queen Anne's War there followed a generation of peace during which the governors of New France built a line of fortified posts: Louisbourg on Cape Breton Island, Chambly on the Richelieu River, Carillon (Ticonderoga) on the portage from Lake George to Lake Champlain; the trading posts of Niagara, Toronto, Detroit, and Michilimackinac extended the line to the west. At the same time, French priests and military emissaries kept the Acadians and the Indian allies of New France aware of their former ties with New France. The Acadians, claiming to be neutral, obstinately refused to take the oath of allegiance to the British crown.

For New France the early 18th century was a period of steady growth. French *défrichements* ("clearings") spread along the St. Lawrence between Quebec and Montreal; the iron forges at Saint-Maurice produced iron for Quebec stoves and even cannons; shipbuilding flourished. The colony nevertheless remained largely dependent on the fur trade, which, in turn, depended on keeping the west open. Access to the far west was frustrated, however, by the three Fox wars (1714–42), in which that tribe strove to close the Wisconsin portages to French traders. Then Pierre Gaultier de Varennes, Lord de La Vérendrye, turned the flanks of the Fox and Sioux by proceeding by way of Lake Superior and the Rainy River to the Lake of the Woods and the Red and Saskatchewan River country. There he found a new region for the French fur trade and also cut into the English trade in the area of Hudson Bay and the Hayes River.

King George's War

The expansion of New France in these years was challenged, however, by the outbreak of the War of the Austrian Succession in Europe in 1740. In America the war became known as King George's War (1744–48). Fighting broke out again in Acadia, on Lake Champlain, and among the English and French Indian allies in the country of the Great Lakes and the Ohio valley. It was a confused conflict of raids and reprisals marked by only one action of major significance—the capture of Louisbourg by an expedition from New England. Holding the St. Lawrence River valley, the Great Lakes, and the mouth of the Mississippi River, the French commanded the better strategic position in America. However, the English colonies, if having a less advantageous location, were far wealthier and more populous.

All this was perceived by Roland-Michel Barrin, Marquis de La Galissonière, the exceptionally able governor of New France (1747–49). He declared in a memorandum to the French court that New France must restore its position by a bold advance into the Ohio valley, which theretofore had not been claimed by New France or its Indian allies. His policy was adopted by his successors, and in 1749 Pierre-Joseph Céloron de Blainville led an expedition down the Ohio to claim the valley for France and to confine English colonists and their fur trade to the east of the Alleghenies. The British colonists, from New York to Virginia, immediately felt the threat to their trade, to their expansion, and to their settlement. In 1749 the Ohio Company was formed in London with English and American support, and the fortress of Halifax in Nova Scotia was built to counter the French fort at Louisbourg, which had been restored to New France by the peace of 1748 ending King George's War. In 1753 an American expedition under George Washington was sent to the Forks of the Ohio to make good the English claim.

The French and Indian War

The French and Indian War. The French had also been active on the Ohio and had opened a line of communication from Lake Erie to the Forks. The rivals clashed on the Monongahela, and Washington was forced to surrender and retreat. This clash marked the beginning of the Anglo-French war known in America as the French and Indian War (1754–63) and in Europe as the Seven Years' War (1756–63). England resolutely supported the colonies and sought to recover the Ohio valley by an expedition under General Edward Braddock in 1755. This, too, was bloodily repulsed on the Monongahela, and both sides then reinforced their colonies with regular troops. It was to be a war for empire; how sternly it was to be fought was made clear by the expulsion of the Acadians from Nova Scotia in 1755.

The French in America, cut off by British blockade, fought to hold the long line from the Ohio to the Bay of Fundy, counting on the results in Europe for a victory. Britain, however, while subsidizing and aiding Prussia, its only European ally, sought victory in America and sent across the Atlantic what was, for that century, an overwhelming body of regular troops in order to stiffen the militia of the American colonies. The first victories went to the French, who captured Fort Oswego and Fort William Henry in 1757 and sternly repulsed the British at Fort

Carillon (Fort Ticonderoga) in 1758. Then numbers and more skillful British generalship began to tell. In 1758 the British captured and razed Louisbourg. In 1759 Sir Jeffrey Amherst began a cautious but irresistible advance from Fort William Henry by way of Fort Ticonderoga to Lake Champlain. In the same year an expedition under General James Wolfe sailed up the St. Lawrence and besieged Quebec, which fell to the British after the celebrated Battle of the Plains of Abraham. Sir William Johnson took Niagara, and John Forbes took the Forks of the Ohio. New France was caught in cruelly closing pincers. In 1760 Amherst closed in on Montreal, and New France capitulated. By the terms of the Treaty of Paris in 1763, all French North America east of the Mississippi was ceded to Britain.

The British victory produced three major results: (1) The danger from New France to the American colonies was ended, thus weakening their dependence on Britain. (2) The British (largely Scots with some Americans) took over and expanded the Canadian fur trade. (3) Britain now possessed a colony populated almost wholly by persons of alien descent.

CANADA UNDER BRITISH RULE, 1763–91

The Royal Proclamation of 1763

The Quebec Act. At first New France was to be governed by the Royal Proclamation (Oct. 7, 1763); it declared the territory between the Alleghenies and the Mississippi to be Indian territory and closed to settlement until the Indians there could be subdued. New France was to become the colony of Quebec, with a royal governor who had the authority to call an assembly. However, the 70,000 French inhabitants of Quebec could neither vote nor sit in the assembly by virtue of their Roman Catholicism.

Few British Americans moved to Quebec (there were perhaps 500 migrants in all), and those who did were attracted primarily by the prospect of taking control of the fur trade. Their bourgeois mentality and their repeated demands for the "rights of Englishmen" tended to alienate the landed-gentry-based British officers who administered the colony. Among the latter was General James Murray, who was appointed the colony's first governor in 1763. Murray sympathized with the condition and difficulties of the French and ignored the demands of the recently arrived Protestants, especially the landed gentry called seigneurs, for an assembly, with the result that an agitation by the Protestants led to his recall. He was replaced in 1766 by General Guy Carleton, who was expected in Quebec to carry out the policy of the proclamation. Carleton, however, soon came to see that the colony was certain to be permanently French. He decided that Britain's best course was to forge an alliance with the elites of the former French colony: the seigneurs and the Roman Catholic church.

Carleton returned to England in 1770 to press his new policy for Quebec on the government of Lord North. The trouble the imperial government continued to have with the colonies to the south secured official acceptance of Carleton's policy. The result was the Quebec Act of 1774, by which the province of Quebec was expanded to include all the inland territory to the Ohio and the Mississippi. The new colony was given a governor and an appointed council. A change in the oath taken by officials, eliminating religious references, allowed Roman Catholics to hold political office. French was made an official language, along with English, and French civil law was restored, but with English criminal law. The Roman Catholic faith was recognized, the church was allowed to collect the tithes of the faithful, and the seigneurial system was endorsed.

Toward the French Canadians the act was liberal and conciliatory, but toward the American colonies it was menacing. To them it reestablished to the north and west an area despotically ruled, predominantly French and Roman Catholic, and with an alien form of land tenure.

American invasion of Canada

The influence of the American Revolution. Instead of intimidating the American colonies, the act helped push the Americans to open revolt. Indeed, the first act of the American Continental Congress in 1775 was not to declare independence but to invade Canada. The failure of that invasion ensured that the continent north of the Rio Grande would, on the recognition of American independence, be divided between the Americans and the British.

Not all American colonists had supported the cause of independence, and many had resisted it in arms. At the conclusion of hostilities, these loyalists had to make their peace with the new republic, as by far the greater part did, or go into exile. The refugees, known as United Empire Loyalists, were the object of considerable concern to the British government, which sought to compensate them for their losses and to assist them in establishing new homes. Some went to the United Kingdom, others to the British West Indies, but the majority emigrated to Nova Scotia or Canada. Nova Scotia, which to a great extent had been recently settled by American colonists, had not, except for an ineffectual rising or two, joined the revolting colonies. Overawed by British sea power and by the fortress of Halifax, Nova Scotians at first kept quiet, and later many of them even made fortunes privateering against American commerce. Easily reached by sea from New York, Nova Scotia became the chief refuge of the loyalists. Some settled in the peninsula itself, some in Cape Breton and in the separate colony of Prince Edward Island. A large number, however, settled along the St. John River, north of the Bay of Fundy. Dissatisfied with tardy government from Halifax, they promptly agitated for a government of their own, and equally promptly the new province of New Brunswick was created for them in 1784, with its own governor and assembly.

In Quebec the loyalists simply crossed the new frontier and settled along the St. Lawrence River. Their impact in Quebec was even greater than in Nova Scotia, resulting in a new province and, perhaps more importantly, the Constitutional Act of 1791. The loyalists who settled in Canada were for the most part quite different from those who went to what were soon to be called the Maritime Colonies. The latter had possessed an elite of government officials and professional men, often loyalist regiments with their officers and men, from the long-settled seaboard areas. The Canadian loyalists, however, were largely from upper New York, especially the Mohawk valley country, and were almost wholly simple frontier folk and recent immigrants, driven from their homes by neighbours who often used the Revolution to dispossess them of their lands; hence, the bitter fighting along the frontier and the long loyalist hatred in the new province for all things American.

Their coming transformed the character of the population of Quebec. That province had been given a government much like that of New France, except for the important office of intendant, and the province was in population almost wholly French, as it was in civil law. Most loyalists had one desire, to hold the land granted them in simple ownership, something the civil law of Quebec did not allow. Some of them—how many is uncertain—also wanted representative government, denied by the Quebec Act. Their representations reached London and were listened to with respect, as they were, after all, from people who had lost everything fighting for the crown.

Appeasement of the loyalists

The Constitutional Act of 1791. Their appeals caused a great problem for the British government. The measures taken in the Quebec Act to conciliate the French could not in honour or policy be withdrawn. Yet the loyalists could not be required to live under French civil and land law and without the representative assembly to which they were accustomed. One obvious answer was to divide Quebec into separate French and English provinces. The English province would have, of course, English common law and an assembly. The French province might have been left with the forms of government provided by the Quebec Act. But there had already been one revolution in America, and, by 1789, another had broken out in France. British statesmen felt that the former had occurred partly because Americans had not been granted the British constitution in its proper forms. The thing to do, therefore, was to give not only the new province but Quebec as well the British constitution in its entirety as far as circumstances might permit. The result would be, it was hoped, to assimilate the French population.

After a fiery debate in the British House of Commons, the Constitutional Act of 1791 gave the same constitution to the colonies of Upper and Lower Canada. Nothing that had been given the French in 1774 was revoked, but the form of government was changed to the familiar one of governor with his executive council, a legislative council, and an assembly elected on what was for the time a wide franchise. The result of this last provision was that the first assembly in 1792 had a majority of French members.

NATIONAL GROWTH IN THE EARLY 19TH CENTURY

Population trends. The coming of the loyalists changed the composition of the population of the British North American colonies by adding elements at once American yet profoundly attached to British institutions; it also increased the population by some 6,000 in the old province of Quebec. To these were to be added the unknown numbers of "late loyalists"—settlers, primarily land seekers, who arrived from the northern states as late as 1812. Some 80,000 came to Nova Scotia, although not all remained; of these, some 20,000 settled in what became New Brunswick, and a few hundred on Prince Edward Island.

The ethnic mixture

They added also to the growing diversity of the population of the colonies. In Newfoundland there were already the west country English and a growing number of Irish—a total of 26,505 in 1806. In Nova Scotia there were, in addition to New Englanders, loyalists, and Yorkshiremen, the Germans of Lunenburg and the Highland Scots of Pictou county and of Cape Breton Island—in all, an estimated 65,000 in 1806, with 2,513 on Cape Breton Island. New Brunswick had a population of about 35,000 in 1806, mostly loyalists or of loyalist descent, but already the southern Irish, drawn by the timber trade, were beginning to appear on the rivers of the north shore. Prince Edward Island, with a population of 9,676 in 1806, had some Acadians, some loyalists, some English, Scots, and Irish. In Upper Canada in 1806 the population numbered 70,718; in Lower it was estimated at 250,000 for the same year.

The first Canadian mosaic had taken shape as it was to remain for a century, a population of British, French, and German. The British element was to be steadily reinforced by northern English, coming by way of Liverpool, Highland and Lowland Scots, and southern and northern Irish. The result was the creation of a society in which religious liberty and a great measure of social equality were necessary for social cohesion and common effort.

Until 1815, however, the number of immigrants was small: Highlanders for Glengarry county in Upper Canada, disbanded soldiers in Lanark county south of the Ottawa River, and a straggle of Irish after the crushing of the rebellion of 1798. Nor did the numbers greatly increase after 1815; not until 1830 did the English, Scottish, and Irish begin to come to the British North American colonies in great volume. Afterward, thousands came each year. The British North American colonies became predominantly British in population, except in Lower Canada, a fact that was to determine the course of Canadian history for the next 100 years.

The Montreal fur traders. The redivision of the continent begun by the American Revolution had been intensified by rivalry in the fur trade. The French fur trade of Montreal had been taken over by British fur traders who conducted the trade with the aid of French experience and skill. The British supplied the capital, and the French voyageurs supplied the skill of canoeists and the knowledge of the country and the Indian. These "Montrealers" pushed the trade with great boldness southwest from Montreal, where they had persuaded the British government not to surrender the fur posts after 1783 on the ground that debts owing loyalists had not been paid by the United States. Thus, the trade of the lands lost by France in 1763 and by Britain in 1783 was kept tributary to Montreal rather than to New York and Philadelphia. The Montrealers also pushed the trade into the far northwest. By 1784 most of the Montreal traders to that region had combined in the North West Company, which dominated the trade from Montreal and seriously challenged the trade and legal monopoly of the Hudson's Bay Company. In the southwest, however, they were slowly squeezed out by the American government and by John Jacob Astor's American Fur Company. So the British traders pushed ever harder into the farthest northwest. In 1787 one partner, Peter Pond, gathered information about a great river that he was convinced flowed into the Pacific. In 1789 a young Scot, Alexander Mackenzie, followed that river, later named after him, to an outlet in the Arctic Ocean. It was a bitter disappointment, for Mackenzie had hoped to find a passage to the Pacific by which the high cost of bringing furs to Montreal could be avoided. In 1793 Mackenzie crossed what is now British Columbia to the Pacific at Bute Inlet. It was a great feat of exploration but did not lead to a route to the Pacific.

The North West Company

The War of 1812. The War of 1812 was to a large degree caused by the Anglo-U.S. rivalry in the fur trade. British traders and soldiers had supplied Indian tribes and afforded them moral support in their contest with the advancing U.S. frontier. Britain had surrendered the western posts by the Jay Treaty of 1794, but the cause of the Canadian fur trade and of the Indians remained the same—the preservation of the wilderness. Certainly, apart from single ship actions and privateering, the war was fought for the conquest of Canada and its elimination as an ally of the Indian. In the end, the war was a stalemate and closed with no concession by either side. It did, however, push back the Indian frontier, increase the breach between the United States and the British North American colonies, and confirm the U.S.-Canadian boundary.

That boundary had been fixed in 1783, by a line running from the mouth of the St. Croix River to the "high lands" dividing Quebec from Maine; then by the mountains between the St. Lawrence and Connecticut River valleys to the 45th parallel; by that line to the St. Lawrence; then by centre line of the river and the Great Lakes and the Pigeon and Rainy rivers to the northwest angle of the Lake of the Woods. The Treaty of Ghent in 1814 confirmed this, although the location of the Maine–New Brunswick boundary remained in dispute until the Webster–Ashburton Treaty of 1842. A convention in 1818 reduced the rights of U.S. fishermen along the shores of the Atlantic colonies and made the 49th parallel the boundary from the Lake of the Woods to the Rocky Mountains. Beyond, the Oregon Territory was to be jointly occupied for a period of 10 years, an occupation ended by the division of the territory, after some threat of war over the U.S. claim to the whole, by extending the 49th parallel to the Strait of Georgia in the Oregon Treaty of 1846.

The 49th-parallel boundary

The Rebellion of 1837. The colonies grew quietly but steadily after the War of 1812, aided by the development of the timber trade and shipbuilding and by increased immigration from the British Isles. The growth of population and social sophistication led to growth of a reform movement, stimulated by similar movements in Great Britain and the United States. The movement was essentially attempted to make the institutions of the colonies more popular and less the monopoly of the royal governors and the appointed and permanent councils. The rise of a middle class of professional men, especially lawyers and journalists, and the importance of religious liberty in a population so diverse in faith underlay this development. In the Maritimes the movement proceeded peaceably and constitutionally under the leadership of forceful but moderate reformers, such as Joseph Howe of Nova Scotia. But Upper Canada, under the leadership of the irascible Scots democrat William Lyon Mackenzie, was more militant. In Lower Canada, matters became much more serious for two reasons: one was that the political division was to a great extent a national split between English and French; the other was that the wheat farming of the habitants was primitive and dependent on new lands and exhausted both the fertility and the extent of the good soils. An economic depression in 1837 intensified political grievances and burst into open rebellion. A hasty and unplanned republican rising in Lower Canada was crushed by the British regular garrison and the militia.

The Union of Canada. The abortive rebellions, however, did dramatize the need to reform Canada's outmoded and constrictive constitution. The "Canadian question" became a leading issue in British politics. Lord Durham was sent out as governor-general with a royal commission to enquire into the causes of the troubles. Durham's stay in Canada was brief, but his enquiry was sweeping and his recommendations trenchant. Durham perceived that

The Durham Report (1839)

the colonies had stagnated and that, if they were to live side by side with the dynamic United States, they must be brought into the full stream of material progress. One political means to this end was union. Durham decided the time for the union of all the North American colonies had not yet come, but he did recommend the reunion of at least the two Canadas in order to realize the economic possibilities of the St. Lawrence valley. In Durham's view, union would also hasten the assimilation of the French, whom he saw as a backward people. Second, he adopted a proposal of certain Upper Canadian and Nova Scotian reformers for "responsible government." This would make the colonial executive responsible to the assembly and assure colonial self-government.

The British government refused an explicit grant of responsible government but did accept the proposal to unite the Canadas. In 1841 the United Province of Canada came into being under a new and dynamic governor, Charles Poulett Thomson (later Lord Sydenham). Although the French of Canada East (Lower Canada) outnumbered the English of Canada West (Upper Canada), both sections received an equal number of seats in the new legislature. This was done to facilitate assimilation of the French, but the French, led by such astute reform leaders as Louis Hippolyte LaFontaine, took advantage of divisions among the English-speaking legislators. They allied themselves with the reformers from Canada West to push for responsible government and to make themselves indispensable for governmental stability. In Britain the success of the Industrial Revolution led to the growth of free-trade liberalism and a desire to dismantle the colonial empire. The last British tariffs (the Corn Laws) were repealed in 1846, and colonial governors were instructed to implement a policy of responsible government. The policy received its first real test the following year, when the reform ministry headed by LaFontaine and Robert Baldwin of Canada West passed a law to compensate victims of the 1837 rebellions. Governor Elgin signed the law despite strong opposition from conservatives. In reaction a mob burned the parliament buildings in Montreal.

Achievement of responsible government

The British North American colonies, then, had achieved self-government by 1848; the next decade saw their laws and institutions remodeled to fit the individual needs of each colony. But, for Canada at least, the time was at hand for expansion. The British repeal of the Corn Laws had deprived the colonies of imperial protective tariffs. The Grand Trunk Railway, begun in 1853, was an attempt to draw the trade of the American Middle West down the St. Lawrence. The Reciprocity Treaty of 1854 in part replaced imperial with continental trade, and the colonies boomed. Economic growth was especially stimulated after 1862 by the American Civil War. When the U.S. government gave notice in 1864 that it wished to abrogate the treaty by 1865, it stimulated colonial politicians to seek unification of the British North American colonies to provide a substitute market. This move was also made necessary by a continuing political deadlock between conservatives and reformers in Canada, by growing fears of U.S. military power, and by a desire to annex the northwest. After the merger of the North West Company and the Hudson's Bay Company in 1821—a merger made necessary by their bloody rivalry—direct links between Canada and the west had been cut. In Canada West, however, a shortage of good agricultural land was forcing young men to leave for the United States to homestead, and demands grew to annex the northwest to provide room for expansion.

The first significant step toward union, later called confederation, was taken in Canada with the formation of the Great Coalition, a government that united Reform Party leader George Brown of Canada West with the Conservatives' John A. Macdonald of Canada West and George Étienne Cartier of Canada East. In September 1864 they attended a conference at Charlottetown, P.E.I., in which Maritime political leaders discussed Maritime union. They persuaded the Maritimes to postpone Maritime union and discuss confederation. At a conference called at Quebec (Oct. 10, 1864), an agreement was quickly reached on a general federal union; this agreement was immediately approved by the British government, which was eager to set the colonies up on their own and to be rid of its obligation to defend them inland from Quebec. There were hitches and failures; New Brunswick voted against union in 1865, then reversed itself in 1866; Prince Edward Island refused to enter until 1873; Newfoundland also refused and did not join Canada until 1949. But the Canadas and the British government applied quiet but strong pressure on the reluctant colonies. In 1867 the three colonies of Nova Scotia, New Brunswick, and the Canadas (now the provinces of Quebec and Ontario) were united under the British North America Act as the Dominion of Canada.

That act, with certain amendments, was the "constitution" of Canada until 1982. It was, in fact, retitled the Constitution Act of 1867 (pairing it in name with the Constitution Act of 1982). The act of 1867 provided constitutions, based on the British model, for the new provinces of Quebec and Ontario, confirmed the language and legal rights of the French, and above all provided for the division of power between the federal government and the provinces. The union was not, at its beginning, a truly federal one because the central government was given broad powers, not unlike those the British government had possessed over the colonies. In time, judicial interpretation and the growth of provincial rights made the union more truly federal. For the moment, however, a strong central government was needed in order to develop the northwest and the colony of British Columbia beyond and to build a railway to the Pacific that would bind the vast new territories to the original Dominion.

Canada from 1867 to 1920

THE DOMINION OF CANADA, 1867–1914

Purchase of Hudson's Bay Company territories

Section 146 of the British North America Act provided for the admission of Rupert's Land (the Hudson's Bay territory) to the new dominion. The first action taken by the federal government was to buy out the title of the Hudson's Bay Company. The negotiations for this purpose, assisted by the British government, were completed in the winter of 1868–69. Canada was to pay the company £300,000 for its title, and the company was to retain 5 percent of the Fertile Belt (land fit for agricultural settlement) and designated areas around its various posts. The Canadian government passed a provisional act for the government of the Northwest Territories, sent out a survey party to begin a land survey before settlement began, and appointed as governor William McDougall of Ontario.

The first Riel rebellion. The government unwisely regarded the acquisition of the northwest as a transaction in real estate with the Hudson's Bay Company. But the company was not the only power in the territory. There was the colony of Red River, and also the métis, who made up more than half the colony. The progeny of white fur-trader fathers and Indian mothers, the métis community was divided between English-speaking Protestant and French-speaking Roman Catholic elements. They had a semimilitary, semipolitical organization of their own. They had long thought of themselves as a new nation, neither European nor Indian but both, sharing the rights of one and the title to the land of the other. And behind them were the powerful Plains tribes—Plains Cree and the Blackfoot confederacy, buffalo hunters not under the influence of the Hudson's Bay Company. Canada had taken no account of the métis or Indians in effecting the transfer. It assumed it could take over from the company and then consider what should be done.

This policy was rendered impossible by Louis Riel, a métis educated in Montreal, who organized resistance in Red River to a transfer to Canada without the people of the northwest having a voice in the transaction. Riel, with the support of armed métis, seized control of Red River and forced Canada to postpone the transfer and to negotiate. The result was the creation in 1870 of the small province of Manitoba, with equality of the English and French languages and an educational system like Quebec's two systems of public confessional schools, Roman Catholic and Protestant. The implication was that the northwest was to be open to French institutions and language, as well as English.

That assumption was to be thwarted, however, by the extreme smallness of the new province, which amounted to little more than the Red River Settlement, and by the dominion's control of natural resources and of the still vast North West Territory. Riel's obstructionism did not block Canada's march to the west, and the dominion at once opened negotiations with delegates from British Columbia. That colony consisted of Vancouver Island (organized as a colony in 1849) and the mainland to the watershed of the Rockies, the latter having first been made a separate colony in 1858, when the gold rush along the Fraser River began, and united with Vancouver Island in 1866. The chief needs of the new colony were responsible government and connection with the east. Union with Canada might afford both, and in the negotiations the chief Canadian representative, George Étienne Cartier, promised both and more—in fact, a railway was to be begun in 2 years and finished in 10 (1881). Faced with such generosity, the Legislative Council of British Columbia could only accept. British Columbia became a Canadian province in 1871.

Addition of British Columbia

The transcontinental railroad. With the addition of British Columbia, Canada extended from Atlantic to Pacific. To maintain that area, however, and to ensure its independence from the United States, it was necessary to build a railway to the west coast. The effort to organize a company to undertake this enterprise, much greater than any railway yet built, was made in 1872. But Sir John Macdonald's government, charged with corruption in its dealing with the head of the new company, fell on the eve of the great depression of 1873. The railway thereafter could only be built piecemeal until Macdonald returned to power in 1878. An economically revived Canada, fortified with a new National Policy of tariff protection, incorporated the Canadian Pacific Railway in 1881, and the line was pushed ahead rapidly with government grants of land and money.

Even so, the railroad came to need new loans from Parliament, and its funds ran out as depression returned. Had it not been for the Riel Rebellion of 1885, which showed the need for the railway in moving troops, the last loan needed might have been refused. Despite the victory in the creation of Manitoba, many of the mêtis—finding life impossible with the influx of new settlers—sold their lands and trekked westward to the Saskatchewan River. Even there, they were followed by the government land survey. The buffalo herds were vanishing, and the railway would supersede transport by boat and cart, from which many of them earned their living. The Plains Indians, alarmed by the depletion of the buffalo and unhappy with the government's treaties, were also restless. The mêtis again organized to claim their rights as they saw them and sent for Riel, living in exile in Montana Territory, U.S. Riel returned and a new armed resistance was formed. Canada had to rush militia to the northwest. Here the new railway, though not quite completed, proved its worth. The troops were rushed out, the rebellion was suppressed, and the railway obtained the grant that enabled it to complete its track across the Canadian Shield and the Rockies. Riel, with several associates, was tried and, despite evidence that he was of unsound mind, convicted of treason, though with a recommendation for mercy. Macdonald, as minister of justice and prime minister, refused clemency. The last spike of the Pacific railway was driven on Nov. 7, 1885, nine days before Riel was hanged at Regina.

Reaction of Quebec

Canada had united its new territories with its old, but there was a fierce reaction in the French province of Quebec. Riel, who had not gained much French sympathy in 1870, was now viewed by nationalist French Canadians as a martyr to the cause of French Canadian rights. The result was the election by a narrow majority of a clerical-nationalist government in Quebec. This produced a reaction in Protestant Ontario, which, in turn, led in 1890 to the abolition of the confessional schools in Manitoba, where the Roman Catholic schools were almost wholly French-speaking. French Canadians thereafter fell back on the provincial rights of Quebec to maintain the rights of French Canadians—a reaction with serious consequences for the Canadian federation. The Conservative Party, which had lost Macdonald by death in 1891, fell from power in 1896 largely because of the Manitoba Schools Question. The Liberal Party, under the French Canadian Wilfrid Laurier, came to power by virtue of a large majority in Quebec. Canada, it seemed, was not to be governed without the support of Quebec, even though the west retained only traces (in the Northwest Territories) of French population and French rights.

The Klondike gold rush. In 1896 gold nuggets were found on Klondike Creek, a tributary of the Yukon River. A rush began in 1897 and swelled in 1898, as miners and adventurers poured in, mainly from the United States. The Klondike was, in fact, the most publicized of all the great rushes. It was the last of the great placer rushes and excited a world weary of depression with stories of the long climb up the Chilkoot Pass and of red-coated Northwest Mounted Police keeping law and order on the gold-rush frontier. Klondike gold, however, was probably the least important mineral discovery of this period. Far more significant for Canada's economy were the copper, lead, zinc, and silver deposits in the Kootenay region of southeastern British Columbia; the coal deposits of the Crowsnest Pass (bordering British Columbia and Alberta); and the gold, nickel, and silver beds of northern and northeastern Ontario and northwestern Quebec. These discoveries stimulated railway and town construction and brought thousands of permanent residents. In the decades that followed, prospectors traced the rich mineral deposits of the Canadian Shield westward from Ontario and Quebec, making major base metals discoveries as well as gold and silver, at Flin Flon, Man., in 1915 and finding rich deposits of radium at Great Bear Lake in 1930. By the 1930s, Canada had become a major mining country.

Increased agricultural base

The land rush in the west. At the same time, the land rush to the prairies widened the country's agricultural base by the settlement of Manitoba and the Northwest Territories. Their population rose from 419,512 in 1901 to 1,322,709 in 1911. Manitoba had already been enlarged westward and northward in 1881. The territories, governed by a governor and appointed council since 1876, had had elected members added to the council and began the traditional Canadian struggle, first for representative and then for responsible government. The latter, however, could only come with provincehood, and the demand arose for the creation of a province between Manitoba and the Rockies. The federal government responded after the election of 1904, and in 1905 not one but two provinces were created, Alberta and Saskatchewan. The provinces were roughly equal in area and extended north to the 60th parallel. A fierce political struggle arose over the question of French schools in the new provinces, and again the cause of French rights outside Quebec received a setback. There were to be separate schools in the Canadian sense of denominational schools supported by the taxes of members of a denomination—Roman Catholics in this case—but nothing like the dual confessional school system of Manitoba before 1890, or of Quebec. Once again the development of the west had disturbed the relations of English and French in Canada.

The Laurier era. For 15 years, Laurier's Liberal government reflected the acquiescent politics of prosperity and progress. Canada seemed at last to be entering on its own century, as the United States had done in the one just past. Nothing better exemplified the confident, easy mood of the Laurier years in Canada than the vast and extravagant railway expansion in response to the settlement of the west and the initial development of the mineral and forest wealth of the nearer, or middle, north. The Laurier government built one transcontinental railway from Quebec to a point east of Winnipeg; from there to Prince Rupert a well-subsidized Grand Trunk Railway of eastern Canada built a subsidiary line, the Grand Trunk Pacific. Not to be deterred by two transcontinental railways in a country that was yet little more than a narrow corridor from east to west, two Canadian private entrepreneurs, geniuses of their kind, William Mackenzie and D.D. Mann, built or bought the Canadian Northern bit by bit with lavish subsidies from provincial governments. By 1914 Canada had one long established coast-to-coast railway (the Canadian

Pacific) and two railway lines from Montreal to the Pacific toiling to complete their tracks in the Rocky Mountains. In such a wealth of easy capital and easy prosperity, governments were not likely to be defeated.

Yet two factors—one as old as Canada, one relatively new—were to disturb the smooth current of prosperity. The former was the never-quite-settled place of the French in an English Canada. The question had flared up in the creation of the new provinces in the west; it now arose over participation in Britain's wars, first the South African War in 1899, then World War I. The result was the growth of a new nationalist movement among French Canadian clerics and intellectuals, who made their voices heard in a new paper, *Le Devoir.* Their spokesman, insofar as a man so independent could speak for anyone but himself, was Henri Bourassa. The second factor was the impingement of the world on a Canada intensely absorbed in its own development and its own troubles. The two were to combine to end the Laurier regime and bring Canada, still troubled, into the world at large.

Foreign relations. Canada's contacts with the world in 1900 were almost wholly through Great Britain and the United States. Indeed, Canada's formal relations with other countries were conducted only through the British Foreign Office because Canada, as a colony, had no diplomatic status.

Relations with Britain

The dependence on Great Britain raised, after 1895, the question whether Canada might be expected, on its own decision, to take some part in Britain's imperial wars. The British colonial secretary Joseph Chamberlain was anxious that the dominion should at least be committed in principle to supporting the mother country. Laurier, at the colonial conference of 1897, remained silent on the issue; thereafter his stand was that the Canadian Parliament must decide whether or not Canada would take any action. When the South African War broke out in 1899, many English Canadians actively urged participation; but some French Canadians, led by Henri Bourassa, were actively opposed. A compromise was reached, by which Canada sent volunteers to serve under British command and with British pay, but the rift between French and English Canadians had been further exacerbated. Also, Britain's naval competition with Germany made Britain eager to have colonial help, preferably by contributions in money or by the colonies' assuming their own naval defense. Again, Laurier sought a compromise. In 1910 he established a Canadian navy, which in time of war, however, was to be placed under British command. The measure was bitterly opposed by the nationalists in Quebec, who argued that conscription in Britain's army would follow. Their clamorous opposition led to the defeat of the government candidate in a Quebec by-election, foreshadowing Laurier's fall from power in 1911.

Canada's relations with the United States were close and informed, but in the final analysis they were relations between states foreign to one another. There had been a long record of border disputes, the settlements of which were resented, rightly or wrongly, by Canadians. There were the perennial difficulties over fishing rights in the North Atlantic and the dispute in the 1890s over the sealing industry in the Pacific. The Fenian raids at the time of confederation symbolized another cause of strain, the Irish-American hatred of England and suspicion of Canada as a British colony. Matters were brought to a head in the dispute over the Alaskan panhandle boundary. The line laid down by treaty between Great Britain and Russia had not since 1867 been marked on the ground by the United States and Canada. It became an urgent issue in 1897 with the Klondike gold rush, as the principal access to the gold fields was through the panhandle, and the disputed territory might contain gold. Canada claimed a line that would have put the heads of major inlets in Canadian territory and so have given it, especially the port of Vancouver, free access to the Yukon Territory. The United States claimed a boundary that would have excluded Canada from the sea. A joint commission of Americans, British, and Canadians found in favour of almost the whole of the American claim, the one British jurist voting with the three Americans. The decision was bitterly resented in Canada, the action of the British member, Lord Alverstone, even more than that of the Americans. In fact, Canada's case was exceedingly weak, but the combination against it of the two powers with which alone Canada dealt made the country realize it must be prepared to look out for itself. Canadian nationalism in a new sense came into being.

U.S.-Canadian friction

Two results followed. One, neither quick nor dramatic but significant, was the creation of the Department of External Affairs in 1909, in preparation for Canadian handling of its own foreign affairs. The other was the cultivation of direct relations with the United States. From this followed the final settlement in 1910 of the long-vexed Atlantic fisheries issues and the creation of the Permanent Joint Commission on Boundary Waters in the previous year.

Relations between Canada and the United States were assuming a new guise, and with reason. The United States was beginning to turn to Canada as an outlet for investment and as a source of raw materials, particularly minerals and newsprint. An exchange of ideas began on a new scale, particularly in the ideas of the Progressive movement, which advocated a wide range of reforms to combat the growing social evils caused by industrialism. These ideas were influential on both sides of the border, in Canada sometimes more than in the United States, as in the creation of the publicly owned and operated Ontario Hydro-Electric Power Commission in 1906. To some degree these developments were upset by the Canadian election of 1911, when the Conservative Party under Sir Robert Laird Borden defeated the Laurier Liberals. Two issues largely dominated that campaign: Laurier's naval policy, which was stimulated by Britain's defense needs in Europe, and a reciprocal trade agreement with the United States. In Quebec the naval policy was denounced as imperialistic. Borden, backed by the business community and renegade Liberals, attacked the reciprocal trade agreement as a sellout of Canada's British birthright and won a convincing victory. This, however, by no means interrupted the growth of the Canadian-American relationship.

Accession of Sir Robert Laird Borden

Although burdened by demands for the distribution of patronage, Borden tried to institute more progressive policies after taking office. But foreign policy issues and defense questions dominated the first years of his government. He struggled to establish a policy of direct cash aid for Britain's naval building program in return for a voice in imperial policies that affected Canada. He was defeated by the Liberal-dominated Senate and rebuffed by the British. When World War I broke out in August 1914, Canada was almost totally unprepared.

CANADA IN WORLD WAR I

At the outbreak of World War I, Minister of Militia and Defence Sir Samuel Hughes scrapped the carefully laid plans for a mobilization of the existing militia and, instead, launched a direct appeal to the men of Canada. Canada was just emerging from a deep recession, and tens of thousands of British-born young men with no work and with patriotism in their hearts rushed to the colours. An initial contingent of 33,000 men sailed for England in October 1914 to lay the foundation for the creation of the First Canadian Division. In April 1915 the Canadians saw their first major action in the Second Battle of Ypres, where they were among the first Allied troops to be gassed. As more volunteers came forward, Borden increased the authorized force levels. By the spring of 1917, four Canadian divisions, constituting the Canadian Corps, were in the field, with a fifth division in England. The entire corps fought together for the first time in April 1917, when it distinguished itself in the Battle of Vimy Ridge in northern France. This force not only earned an enviable record as a fighting force; it was the first authentic expression of Canada in the world. Its strength and reputation meant that Canada could not be treated as a mere colony. The cost to Canada was high. Out of approximately 625,000 who served, about 60,000 were killed in action and another 173,000 were wounded. The Canadian Corps won many battle honours, but its role in history was to ensure Canada that "voice" in imperial and international affairs for which Borden had asked in vain in 1912.

The Canadian war effort

At home, the war effort was scarcely less impressive. Canadian foodstuffs and raw materials were of first importance in maintaining the western Allies. No less important were the millions of shells turned out by Canadian factories. In fact, the war was a significant step forward for Canadian industry, which had to learn complicated mass production techniques and apply them to the manufacture of everything from wooden shell crates to training aircraft. The rapid growth of the munitions industry created an acute labour shortage that brought increased female participation in the industrial work force. It also promoted the growth of labour unions. At the same time, the accelerated demands of the war economy brought high inflation, which the government was unable to control despite increasingly interventionist policies. Strikes and lockouts grew to crisis proportions by the last year of the war.

At the start of the war, Borden had envisaged an essentially voluntary war effort: employers were urged to treat their workers fairly, workers were urged to curb wage demands, producers were urged to keep price increases down, and men were urged to enlist. As the war dragged on, more and more English Canadians began to view the war as a Canadian national war effort, not simply as another British war in which Canadians were taking part. By 1917 the government was trying to regulate many facets of Canadian economic life. It nationalized bankrupt railways. It introduced income taxes. It controlled some commodity prices. And in the spring of 1917 it introduced compulsory military service—conscription—in response to a growing manpower crisis in the Canadian army. Conscription tore Canada apart. French Canada had never been enthusiastic about the war, and many fewer French Canadians volunteered for military service than did English Canadians. To make matters worse, French nationalist feeling had been reawakened by new troubles with respect to the use of the French language in schools in French districts in Ontario and Manitoba. French Canada, led by Laurier, opposed conscription but was overridden by the formation of a Union government—almost wholly English in personnel—and in the wartime election of 1917. But Canada was divided as it had not been since 1837.

Despite this rift at home, the entry of Canada into the world of nations went ahead. In 1917 the British government under Prime Minister Lloyd George formed an Imperial War Cabinet, of which the prime ministers of the dominions were members, to conduct the war and to plan the peace. In reality, if not yet in name, the British Commonwealth of Nations had come into being. This was recognized by Article IX of the Imperial War Cabinet in 1917, which stated that the British Empire was made up of self-governing nations as well as colonies, with India in a special position. Henceforth, it was hoped, a common policy would be worked out by government conferences in peace as well as war.

Canada since 1920

CANADA BETWEEN THE WARS

The labour–farmer revolt

Turmoil at home. During the war discontent had increased in virtually every region of Canada and in almost all of its social classes. When the fighting ended, patriotic constraints on demands for change disappeared, and organized labour and the farmers mounted a revolt that swept across Canada. In 1919 the Conservative government of Ontario was turned out by a farmer–labour alliance led by the United Farmers of Ontario. United Farmers governments were elected shortly afterward in Alberta (1921) and Manitoba (1922). In federal politics the agrarian-based Progressive Party in the 1921 election became the second largest party in the House of Commons. The agrarian revolt was marked by demands for farm price supports and regulation of the grain and transportation industries. At heart, however, it was aimed at curtailing the growth of the power of the cities.

A labour revolt paralleled the uprising on the farms. The virtual doubling of union membership across Canada during the war and the failure of the Borden government to control inflation stimulated militancy. There was an upsurge of industrial unrest despite governmental efforts to impose peace. In 1919 the unrest peaked with a six-week general strike that paralyzed Winnipeg and sparked sympathetic strikes across Canada. The Winnipeg General Strike was crushed by a federal government gripped by a hysterical fear of revolution. By 1921 the labour revolt had subsided, partly because of federal intervention, partly because of the onset of an economic turndown that brought increased unemployment and a virtual collapse of union power.

Commonwealth relations. As a result of their efforts during the war Canada and the other dominion powers demanded separate signatures to the treaties with the defeated countries and won at least the right to sign as members of a British Empire Panel. They also demanded and got, despite the doubts of the United States and France, membership in the newly organized League of Nations. Thus, Canada finally became a full-fledged member of the community of nations.

Canada followed an isolationist policy between World Wars I and II. This was mainly a consequence of the return to government in 1921 of the Liberal Party, now more than ever dependent on French Canadian support. French Canadians were themselves isolationist, and they strengthened the general disposition of Canadians to express their new national feelings by the completion of autonomy in the empire and by resuming their material development as a North American country. The new prime minister, W.L. Mackenzie King, and his government were firmly nationalist and isolationist, as was made very clear by their prompt refusal to back the United Kingdom's policy in Turkey in 1922. This, in effect, ended the hope of a common imperial policy. Instead, there would be conference, consultation, and sharing of information, but freedom of action.

The Balfour Report of 1926

King was primarily motivated by his desire to maintain national unity. He knew that a close relationship with Britain would remind French Canadians of the conscription crisis, and he was determined not to split Canada over questions of foreign policy. He was a willing partner in efforts by the Union of South Africa and the Irish Free State to weaken the formal ties of empire—efforts that culminated in the Balfour Report of equality of status at the 1926 Imperial Conference. That report was embodied in the 1931 Statute of Westminster. The statute ended all legislative supremacy of the Imperial Parliament over the dominion parliaments and made them, when they proclaimed the Act, sovereign states sharing a common crown. The Commonwealth had become a legal reality and Canada an independent nation.

Canada's new independence was reflected by its establishment of its own foreign service. Canadian ministers were appointed to Washington, D.C., in 1927 and to Paris and Tokyo in 1928 and 1929, respectively. (In the United Kingdom and Canada, officers called high commissioners played much the same role after 1928, although the office was to some degree political and not just diplomatic.)

The Great Depression. The quest for international equality was to some degree obscured by the onset of the Great Depression. With an economy largely dependent upon the export of farm products and raw materials, Canada suffered severely during the depression period from 1929 to 1939. The Liberal Party, except for a brief interval in 1926, had been in power since 1921. In the general election of 1930, the Conservative Party, led by R.B. Bennett, succeeded in winning the election—partly because the King government had refused to give federal funds to the provinces to combat unemployment and give relief, and partly because Bennett promised to blunt the effects of the depression by protective tariffs. If the attitude of King was old-fashioned and not up to the needs of the times, so were the election promises of Bennett. The depression was so far-reaching in its effects that only drastic measures could lessen its impact and begin the process of recovery.

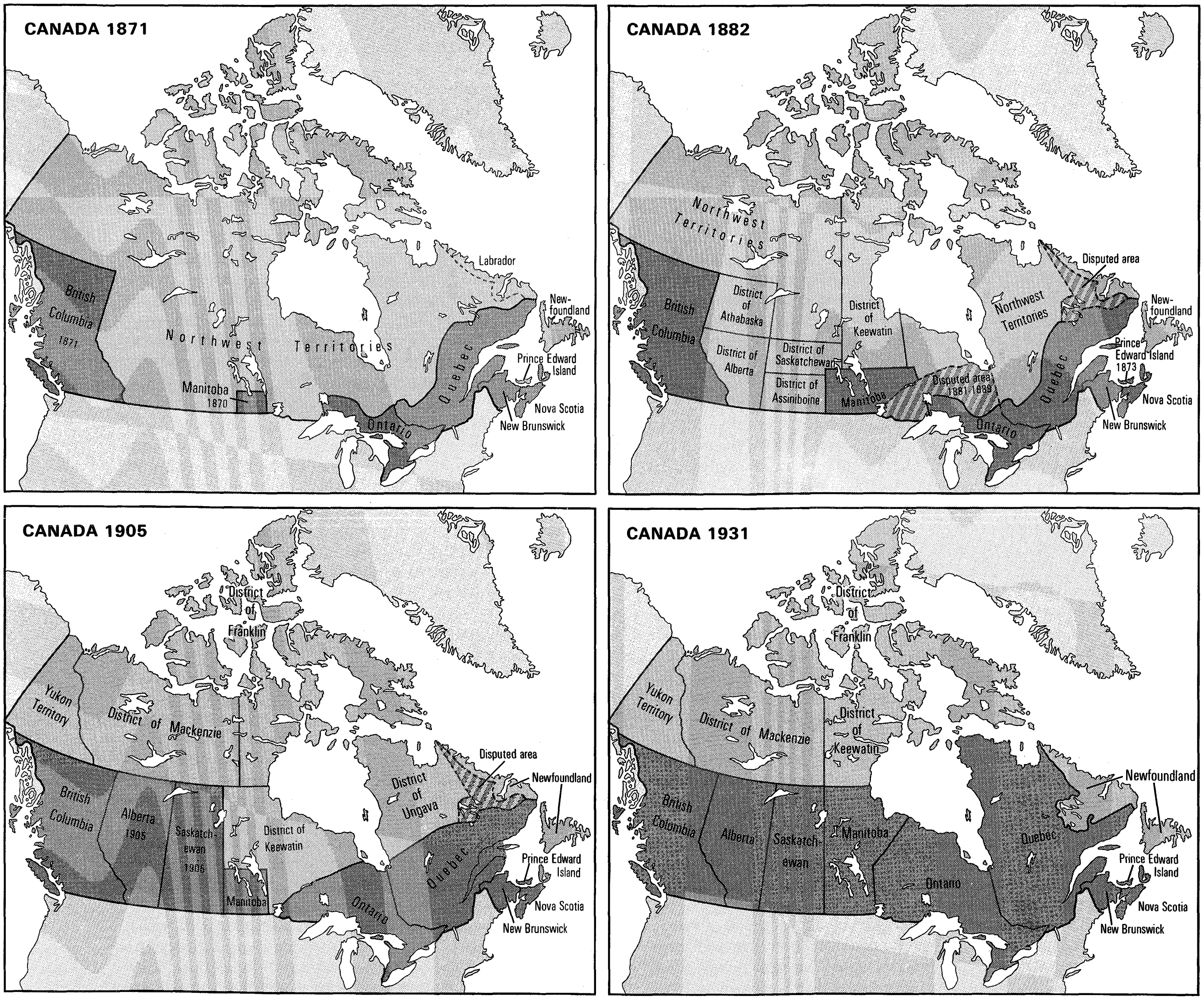
The Bennett government did little to increase the already high tariffs that had been in effect under the Liberals, but it did stiffen the application of customs regulations. The Conservatives also sought relief for Canadian producers by reviving an old dream of Joseph Chamberlain that the

The Imperial Economic Conference

empire should have a common tariff policy. The obstacle created by Britain's policy of free trade had ended with the modification of that policy in the 1920s. Political imperialism had ended, but perhaps commercial imperialism would be revived. Prime Minister Bennett arranged an Imperial Economic Conference at Ottawa in 1932. The national interests of the representatives made agreement difficult, but all countries were desperate. The dominions badly needed the British market for their foodstuffs and other primary products, and Britain itself could not afford to ignore protected markets in the dominions. The Ottawa Agreements were a series of preferences given by Commonwealth countries to one another, to some degree at the expense of producers in other countries, particularly the United States.

The depression, however, merely intensified. Canada had known widespread unemployment in the mid-1870s and in the immediate postwar years. But an economy still predominantly agricultural had largely, if painfully, absorbed the unemployed. In 1921, however, supremacy began to pass from agriculture to industry, and from countryside to city, and the trends thereafter were rapid. The cities had now to provide for mass unemployment, and Canadian municipalities were required by law yet unfitted in financial resources to furnish welfare and relief. The municipalities had therefore to turn to the provinces for aid; the weaker provinces had to turn to the federal treasury. Thus, not only was the economy being dragged down to unexampled levels of prices and production, but the very framework of government was threatened in the poorer provinces. What happened in Newfoundland, where responsible government was replaced by administration by a commission aided with British funds, came close to happening in the prairie provinces. The brighter side of the picture was that the mining industry of Ontario and Quebec kept the revenues of those provinces comparatively good, and so the greater part of the country remained in a condition to maintain the federal revenues and make possible the support of the weaker provinces. Confederation had never been more important.

Prime Minister King might refuse to aid the provinces in 1930, before the second great slump of the depression, but his successor had no such choice. Prime Minister Bennett's own remedies—stiffened protection and imperial preference—could only act slowly at best. But conditions were drastic by 1931, aggravated by unprecedented drought on the prairies. Bennett began, on entering office, to advance federal funds to the provinces for relief works. The provinces had already come to the financial aid of their municipalities, and a relief system was worked out for the unemployed. But all the means of aid that were employed ended with the giving of relief; the government still looked for recovery from the depression. The rise of parties with more fundamental solutions than relief therefore began. One was the Co-operative Commonwealth Federation, a party socialist in theory if pragmatic in fact. Another was Social Credit, a movement urging the payment of social dividends to fill the gap between the costs



Canadian political development, 1871–1931 of production and the cost of purchase. There was even a small Fascist Party in Quebec, as well as a splinter group led by Paul Gouin, a Liberal rebel, which in 1936 joined with Maurice Duplessis's Conservatives to make up the Union Nationale, which became a nationalist "Quebec First" party. Disquieting findings of a Price Spreads Commission and the growing desperation of the farmers led Bennett to listen favourably to praise of the New Deal in the United States. Perhaps personally convinced, certainly politically desperate, in 1935 he had his last Parliament enact a Canadian New Deal. It quite failed to remove the resentment his government had provoked during its years of office, and the Liberals returned to power easily.

The Reciprocity Treaty of 1936

King, despite some advanced ideas in his youth, believed Bennett's New Deal was unconstitutional, as the courts in 1937 found most of it to be. King and his government believed that the way to end the depression was to stimulate international trade. The Ottawa Agreements, of course, remained in force. Canada, however, availed itself of the U.S. policy of more liberal trade agreements, resulting in what was called the Reciprocity Treaty of 1936. Reciprocity and imperial preference were reconciled. In the long run, U.S.-Canadian trade was to grow to enormous proportions, but for the time being the depression continued, as did the drought.

The Liberal government did concern itself with the devastation wrought by the depression on government finances. In 1936 an official inquiry made by the newly created central Bank of Canada revealed that the Prairie Provinces were at the end of their economic rope. A distinguished Royal Commission was set up to inquire into the allocation of revenues between the federal and provincial governments. The resulting report in fact amounted to a comprehensive study of the constitutional and financial development of government in Canada, and of how the depression had revealed its weaknesses. The federal government, with unlimited power to tax, lacked the power to spend on needed issues; the provinces possessed the necessary constitutional power but, except for Ontario, their financial resources were inadequate. The commission in 1940 recommended the assumption by the federal government of provincial debts, a scheme of federal unemployment insurance, and a reallocation of revenues between the two levels of government, on the principle that all provinces had the means to maintain a national level of governmental and social services. The first two were done, relieving the credit of the provinces and strengthening the federal government, but on the latter there was no agreement, as it involved a redistribution of income between wealthier central Canada and the Maritimes and the Prairies. At that time both Ontario and Quebec were strongly provincialist and resisted redistribution.

Canada's policy toward Hitler

Growing international tension. Even domestic distress was to some degree submerged in the second half of the 1930s by the worsening outlook in international affairs. The external interests of Canada shifted from the development of the Commonwealth to the fate of the League of Nations and the first shocks of aggression in Asia and Europe. Canada was too preoccupied with its own affairs up to 1935 to take great note of Japanese aggression in Manchuria or the rise of Hitler in Europe. By that date, however, the fate of the League of Nations, clearly threatened by unprovoked aggression, drew more and more attention. From 1936, King supported the French and British policy of appeasement, refused to make any public commitments to aid Britain in the event of war, and declared that Parliament would decide Canada's course if and when fighting broke out. He did this even though he knew that strong ties of culture, emotion, and nationality still bound most English Canadians to Britain and that these ties would inevitably bring Canada into war on Britain's side. He said as much to Hitler in a visit to Germany in 1937. His plan, however, was to put commitment off to the last possible moment so as not to alienate French Canada until war had actually begun.

CANADA IN WORLD WAR II

When the Parliament of Canada voted on the declaration of war against Germany on Sept. 9, 1939, only a very few members voted in opposition. This unanimity, however, rested on the assumption that there was to be a "limited liability" war effort that would consist mainly of the supply of raw materials, foodstuffs, and munitions and the training of Commonwealth air crew, mainly for the Royal Air Force. Canadian males were to be actively discouraged from serving in the infantry, which was expected to take high casualties, and few infantry units were planned. If this plan were followed, King and other government leaders reasoned, conscription would be unnecessary. King and the leader of the Conservative Opposition had both pledged themselves to a "no conscription" policy even before the war began.

Increased Canadian involvement in the war

The expulsion of the British from Europe and the fall of France in May and June of 1940 totally changed the circumstances. Canada's overseas allies had fallen or were in danger of doing so, and it immediately concluded an agreement at Ogdensburg with the United States for the defense of North America. Moreover, Canada now stood in the forefront of the war. After England, it was now the second most powerful of Germany's adversaries. The emphasis on supply had to give way to an emphasis on combat forces. Necessarily, the question of the allocation of manpower, including compulsory service overseas, came to the fore and was made more urgent by the Japanese attack on Pearl Harbor in December 1941. King, resolved at almost all costs to avoid a repetition of the Anglo-French cleavage, had to recognize the changed situation by holding a national plebiscite in 1942 to free him from his pledge and allow conscription if necessary. It was strongly carried in English Canada but heavily defeated in Quebec. Thereafter, compulsory service for home defense was enforced, while overseas service remained voluntary until the last year of the war. Then in 1944 unexpectedly great casualties in Normandy led to a "conscription crisis" and the ordering of 16,000 home defense men overseas. The gulf of 1917 between English and French, reopened in 1942, was deepened.

Not only was Canada's war effort in World War II far more massive than that in World War I, it also had a much more lasting impact on Canadian society. By the end of the war, more than 1,000,000 Canadians (about 50,000 of whom were women) had served in the three services. Casualties were much lower than in the previous war, with approximately 42,000 killed or having died in service and 54,400 wounded. The Royal Canadian Air Force provided an entire group (6th Group) for the Royal Air Force's Bomber Command, and the five Canadian divisions were united in the First Canadian Army under Lieutenant-General H.D.G. Crerar. Canadians saw action in battles as far afield as Hong Kong, in December 1941, and the invasion of Normandy, in June 1944, when Canada was assigned one of the five invasion beaches.

The domestic war effort was no less significant. Canada hosted, and paid much of the cost of, the British Commonwealth Air Training Plan, which trained more than 100,000 Commonwealth air crew. Canadian factories turned out everything from rifles to Lancaster heavy bombers, and Canadian scientists, technicians, and engineers worked on the latest weapons technology, including the atomic bomb, for which Canada supplied the uranium ore. Canadian foods, direct cash contributions to Britain, and munitions for the Allies, including the Soviet Union, contributed to the overall war effort.

Unlike the situation during World War I, the government intervened in almost all aspects of Canadian life to regulate the war effort, ensure a smooth flow of men and materials, and curtail inflation. Agencies such as the Wartime Prices and Trade Board and the National War Labour Board marked a massive growth in the federal government that brought a surge of government spending and a vast increase in the civil service. Toward the end of the war the King government launched Canada into the era of social welfare spending and Keynesian economics with a major veterans' benefits program, family allowances, farm price supports, compulsory collective bargaining, and a national housing program. It would undoubtedly have gone even further than it did in 1945 and 1946—it was considering a national health insurance plan—but for the opposition of provincial governments headed by Ontario and Quebec. Despite that opposition, however, the war produced a significant shift of power toward Ottawa.

POSTWAR CANADA

U.S. and British relations

International commitments. The most significant outcome of the war for Canada in its foreign relations was the relative decline of Britain and the emergence of the United States as the world's foremost economic and military power. Canada's relations with Britain remained unchanged; those with the United States became closer. The creation of the Permanent Joint Board on Defence in 1940 was a significant harbinger of that shift. For the first time in its history, Canada coordinated its defense planning with the United States.

Canada's shift in orientation from Britain to the United States did not come all at once and did not progress without hitches. In early 1948, for example, King balked at concluding a free-trade agreement with the Americans. But given Britain's growing economic, political, and military weakness, and the rise of the United States to superpower status, it was inevitable. It posed a major dilemma for Canada, especially in the Cold War era: how to safeguard Canadian sovereignty while integrating Canada more and more into the United States' economic, diplomatic, and military spheres of influence, especially when Canada shared to a considerable degree the U.S. view of the postwar world. One answer was to avoid bilateral arrangements with the Americans where possible and to involve Canada in multilateral organizations such as the United Nations, where U.S. influence would be somewhat diffused. Most informed Canadians welcomed the United Nations, and the Canadian government took a vigorous part in its creation. But King, mindful of his own lifetime battle to remove Canada from the trammels of British imperialism, was dubious of a world to be dominated by the great powers. His advisers, for their part, wanted to find some way for Canada to play a significant role in the world. The government therefore advanced the concept of a status of "middle power"—that is, a state strong economically if not militarily. Members of the United Nations should function appropriately, according to their capacities. The idea in practice meant that Canada should concern itself mainly with economic policy in world affairs and with aid to developing countries.

Similarly sensible was the Canadian decision to use its considerable knowledge of nuclear fission not for military purposes but exclusively for peaceful and economic ones. Canada, in short, wanted—as in 1918—to get back to its original peaceful character and to resume its own development, this time in the world at large. That desire, however, was just a dream in a world increasingly split between East and West, between supporters of democratic capitalism and partisans of communist totalitarianism.

Canada and the Cold War

Although the Cold War was born in Europe, primarily in the failure of the wartime Allies to decide the fate and future of Germany, Canada was involved from the very start when in September 1945 the defecting Soviet cipher clerk Igor Gouzenko revealed extensive Soviet spying operations in Canada and the United States. These revelations, combined with Soviet intransigence at the United Nations and Soviet aggressiveness in central and eastern Europe—particularly the communist coup in Czechoslovakia and the Berlin blockade—convinced Canadian leaders of the malevolent nature of Soviet communism.

As the Cold War intensified, the idea of a regional agreement for the defense of western Europe against Soviet pressure or attack was taken up. Devoted supporters of the United Nations in Canada as elsewhere were dismayed, for they felt that regional agreements militated against the global purposes of the general organization. The Canadian government, nevertheless, supported the United States' proposal for an alliance of North Atlantic powers. Yet Canada insisted that this alliance should not be purely military, but also political and economic. The minister of external affairs, Lester B. Pearson, pressed strongly for the adoption of that principle. It was accepted in Article 2 of the North Atlantic Treaty Organization (NATO) in 1949, sometimes referred to as the "Canadian article." The article, however, has had little notable effect. Whether that is a comment on a Canadian lack of realism or the lack of percipience in others, it remains evidence of the outlook and intent of Canadian foreign policy and of informed Canadian opinion at the time. Canada as a member of NATO also, for the first time in its history, assumed serious peacetime military commitments, maintaining out of its armed forces an infantry brigade and 12 air squadrons and contributing ships to the naval forces of NATO until 1969. Then a change of emphasis in foreign policy led to a reduction, much criticized by its NATO allies, of the Canadian contribution.

The Korean War

Just as NATO was a test of Canada's seriousness in entering world affairs, so, too, was the Korean War of 1950, which also tested Canada's relationship with the United States. When that country decided to assist South Korea in resisting invasion by the forces of North Korea, the question for Canadians was: Was this an endeavour to uphold the peace settlement that had divided Korea at the 38th parallel of latitude, or was it simply part of a U.S. crusade against communism? In the latter, few Canadians had any interest—not even French Canadians, for all their Roman Catholic heritage. The government decided that it was the former, and military and naval contingents from Canada served with the U.S. and UN forces.

More significant was the fact that informed Canadian opinion was already markedly diverging from the opinion of the majority of Americans on the subject of the new communist regime in China and would have approved recognition of it as the government manifestly supported by the Chinese people. But the government, though sympathetic, thought it wise not to challenge U.S. opinion on an issue in which no major Canadian interest was at stake. Most important, from a practical standpoint, was the economic outcome of the Korean War, which ended any possibility of a postwar depression—expected but not experienced from 1945 to 1950—and helped launch Canada on its extraordinary development in the 1950s.

U.S.-Canadian relations. The policy of the Liberal government (in power since 1935), wartime cooperation, and the close interconnections between the two economies had brought the two neighbours into a more intimate relationship than ever before. The United States, with British power and influence ended in Canada, no longer had intermittent designs on that country and had dropped the language of Manifest Destiny. Most Canadians were persuaded that the development of economic relations did not necessarily carry with them political relations that would end in annexation. The two nations, then, accepted and took one another for granted. In that development, however, there were certain dangers. Americans, including the U.S. government, were accused by Canadians of taking Canada for granted in the sense of being unaware of and indifferent to Canadian sentiments and changes. Canadians, in turn, were apt to claim both a special standing with the United States and independence of it. None of this was ultimately of great importance, but two new trends were to prove significant. One was the growth of "continentalism," a special relationship not provided for in the theory of national independence—except perhaps in the joint commissions, such as the Permanent Joint Commission on Boundary Waters and the Ogdensburg Declaration, in which the parties met as legal equals and in which the fundamental assumption was that both, when in agreement, would contribute a proportionate share of their population and wealth to the common undertakings. The second was the unequal rate of economic and technological development, especially after 1950. The United States, leader of the world in industrial capacity and technology, was nearing the limits, subject to strategic reserves, of its natural resources. Canada, within the inner defense orbit of the United States, had many such resources undeveloped and available. The interest of the United States was, therefore, to have assured access to these resources as they were developed, largely with U.S. capital. This, however, tended to keep Canada a primary producer and a country of relatively low employment and low income. Its national development and its hope of educational and cultural development called for the continued growth, un-

The United States and Canadian resources der Canadian law and control, of its secondary industry. Yet its provinces—owners of the natural resources of the country, except for those controlled by the Northwest Territories—were driven by the need to secure revenue and to satisfy the popular demand for development, and thus were eager to alienate their resources to foreign, usually U.S., investors. This disparity of aim made U.S.-Canadian relations, if much better diplomatically than in the days of territorial expansion and boundary settlements, much more subtle and complicated than ever before. Something of this lay behind Lester Pearson's warnings in the mid-1950s that henceforth those relations would be more difficult. It appeared also in the conclusion of a study prepared in 1965 by Livingstone Merchant, U.S. ambassador to Canada, and A.D.P. Heeney, Canadian ambassador to the United States, that U.S.-Canadian relations could best be handled by "the practice of quiet diplomacy"—that is, with a minimum of public discussion.

A third element of difficulty in U.S.-Canadian relations, which became more apparent in the late 1960s and early 1970s, involved the fact that the main thrust of Canadian policy was to support the United Nations, in which Canada might evade subordination to the great powers. It did so by its peacekeeping efforts (Canadian troops, for instance, served with the United Nations in Egypt after the Suez crisis of 1956) and in cultivating the developing countries and those, such as India, that attempted to play a neutral role between the United States and the Soviet Union in the Cold War. To do this, Canada had the advantage of being at once a materially advanced country and also a small nonimperialistic one. But Canada's special relations with the United States continued and expanded. One notable example of joint action was in the Distant Early Warning System against surprise aerial attack, largely based in the Canadian Arctic. Another was the North American Air Defense Command (Norad) arrangement of 1957 for the defense of North America against aerial and missile attacks, in which elements of the Royal Canadian Air Force were placed under U.S. command. How could Canada be at once an influential, because uncommitted, power to the Third World and an ally, even for limited purposes, of the United States? The opinion increased that it could not, and after the election of Pierre Elliott Trudeau as prime minister in 1968 this attitude began to affect Canadian foreign policy. Canada's contribution to NATO and Norad began to be whittled down in the Defence White Paper of 1970. At the same time, relations with the Soviet Union were improved, and in 1970 diplomatic recognition was extended to the People's Republic of China. Concurrently, however, the stress on aid to developing countries and the idea of Canada as an intermediary with the Third World were played down. The emphasis shifted to more domestic aspects of foreign policy: control of pollution in coastal waters, extension of Canadian fishing rights, and especially to confirming Canadian sovereignty in its Arctic islands and their icebound waters, a concern not much in evidence since the early years of the century. Yet in all these changes the special relationships with the United States remained unchanged, rooted in the facts of neighbourhood and common interest.

Recognition of the People's Republic of China

Canada and the Commonwealth. If the special ties with the United States waxed during these years, the historic ones with Great Britain waned. It cannot be said that this was willed or decided by either party. It was the outcome of the developing circumstances. The traditional ties remained—the common crown; the parliamentary system of government; the desire for much the same kind of world; the same pragmatic, unideological temperament and outlook. Cordial relations between the two governments remained, consultation continued. Canada had, however, achieved the kind of world it wished to inhabit, with the creation of the Commonwealth; Britain, charged with the task of liquidating its empire, began to find the Commonwealth a burden and an embarrassment. Even imperial preference failed of its purpose; Canada exported more to Britain and imported more from the United States, while Britain exported less to Canada. Like relations who have moved apart, each knew it could call on the other at need, but each was going its own way with little to call them together. The British phase of Canadian history, which had begun with the great British immigration of the 1830s, seemed to have ended.

Post-World War II Commonwealth membership policy

The reverse was true for the Commonwealth of Nations during the 1950s and '60s. Canada, as one of the principal creators of the Commonwealth in the early 1930s, had a special if largely sentimental interest in it. But the Commonwealth had consisted of self-governing nations, and, thus, those dependent parts of the British Empire, including India, that were not self-governing, or not wholly so, had not been members of the Commonwealth. With the rapid achievement of independence in most parts of the empire after World War II, a process of which Canadians in general approved, the new nations could choose to leave the empire entirely, or to apply for membership in the Commonwealth. Burma (Myanmar) chose the former course, India the latter. But India had also become a republic in 1947. How could a republic be part of an association bound together by allegiance to a common crown? Suddenly the Commonwealth was seen as an association that might bridge the differences of race and culture in freedom as the empire had done by power. It was agreed among the members of the Commonwealth that republics could be members if they chose to accept the king as "head" of the Commonwealth. Canadians, as members of a republican hemisphere, were perhaps particularly happy to agree, and certainly because the proposal fitted well with their desire to cultivate the widest possible associations, especially with the new and developing nations. Canada, perhaps somewhat smugly, saw a possible role for itself as an intermediary between the old members of the Commonwealth and the new.

Canada's relationship with the Third World

That Canada might, in some sense, play such a part was revealed by the Suez crisis of 1956, a great strain for the Commonwealth as well as for world peace. Australia and New Zealand, for example, were disposed to sympathize with the strategic concern of the United Kingdom, while India was dismayed and angered by what it saw as an act of concerted aggression. Canada, led by Lester Pearson, was able to intervene between a wrathful and humiliated United Kingdom and a righteously angry India, enabling the contestants to save face to some extent and preserve the integrity of the Commonwealth.

The Canadian role of disinterested friend was also played in the crisis in Commonwealth affairs precipitated by the apartheid policy of the Union of South Africa. To a multiracial association such as the Commonwealth, South Africa was not only an anomaly but a reproach. Yet a basic rule of the Commonwealth was that of nonintervention in the domestic affairs of members. The issue of South Africa's racial policy came to a head in the Commonwealth Conference of 1960, at which certain members sought to have South Africa expelled. The United Kingdom, Australia, and New Zealand deplored this violation of the rule of nonintervention. Canada again tried to play the role of impartial intermediary, but when that failed, voted for expulsion. In short, Canada's Commonwealth policy was normally that of supporting the nonwhite members, as later in the matter of economic sanctions against the white regime in Rhodesia (now Zimbabwe).

That was a defensible policy, if pursued realistically and for the purposes of mediation, and even if it led to some feeling against Canada in the United Kingdom. But in 1961 another issue was to arise at the Commonwealth Conference in Accra, Ghana, that raised some question of the whole future of the Commonwealth: the possible entry of the United Kingdom into the European Common Market. For it to do so would mean, if not the ending, at least the diminution of the imperial preferences, which since 1932 had given the Commonwealth a material as well as a sentimental basis. The Canadian delegates professed to be appalled by the prospect and spoke strongly against it as a betrayal of the Commonwealth. But in view of Canada's favourable economic position in trade with Britain, this attitude seemed to the British selfish and unacceptable. Britain was, in fact, not to join the market until 1973, by which time Canada, then under a Liberal government, chose merely to wish Britain well and to look out for Canadian trade with the market as best it could.

Britain and the Common Market

But Britain's entry meant that the Commonwealth would be less and less a matter of material ties and more and more one of tradition and sentiment.

Ideally, perhaps, Britain might, like France, have sought to bring its Commonwealth partners into some form of economic association with the market, as it was to do with New Zealand and the African and West Indian members, heavily dependent on the British market for the sale of their primary produce. But industrialized members such as Australia and Canada, with their economic ties with the United States, had relatively little such dependency, and the members of the market were firmly opposed to their becoming even associated members. There was no hope of a special relationship in that respect. Canada therefore would have to stand on its own, economically as well as diplomatically.

Franco-Canadian affairs. That Canada was regarded in Europe as being at once on its own and also an economic, if not a military, dependency of the United States was revealed by the course of Franco-Canadian relations in the 1960s. France was, of course, as much as Britain a mother country of Canada. Yet practically nothing to reveal that fact had happened since the cession of New France in 1763. Relations between France and French Canadians had been strained by the Seven Years' War. The British regime was watchful against all but private correspondence. Then the French Revolution, particularly its attack on the Roman Catholic church, caused further friction between France and French Canadians. What influence there was was private and literary. There had been readers of the philosophes in New France, and in Quebec French books and ideas were always sure of at least a small audience. Nothing was done officially and formally, however, until the 1850s, when the Second Empire of Napoleon III, Catholic and expansive, took note of French Canada. Not only did the frigate *La Capricieuse* visit Quebec in 1855 but, in 1859, a French consul general was appointed to Quebec, and a form of diplomatic relations began (and continued thereafter). Little more came of this rapprochement. French Canada's true ties abroad were with Rome, not with Paris.

The "Quiet Revolution"

All this changed a century later, particularly after the "Quiet Revolution" began in the province of Quebec with the election of a Liberal government led by Jean Lesage. All changed because French Quebec had wholly changed.

One result was to alter decisively relations between French Canada and France. The former suddenly found an interest in French history, French ideas, and the place of France and the French language in the world. French Canadian students went to French universities, teachers were exchanged, and some liaison developed between the press of the two countries. All this the Department of External Affairs both noted and welcomed. Canada and France had exchanged ministers in 1928 but as between foreign countries.

A warming of established relations proved to be by no means enough, however, either in Quebec or in France. Perhaps the main object of the Quiet Revolution was to arouse and make flourish French culture and the French spirit in Quebec. To do so was not a matter of diplomacy, as generally practiced, or of commerce, but of cultural affairs, for which Quebec had already set up a government ministry. The urgent spirit of the new Quebec was profoundly dissatisfied with the way Ottawa dealt with such matters in the Canadian embassy in France. Quebec began to establish what might be called quasi-diplomatic relations between France and Quebec and had constitutional grounds for thinking it might do so. The federal government could not make treaties with foreign powers on subjects coming under the jurisdiction of the provinces. Cultural affairs, Quebec claimed, were educational, a provincial matter, and therefore Ottawa could not conclude treaties in cultural matters. It followed that Quebec ought to be free to develop its own cultural relations with France and, indeed, all Francophone countries. The matter was difficult and is of continuing concern in Quebec.

De Gaulle's 1967 visit

The second result arose from the personality of President Charles de Gaulle of France. De Gaulle made it his particular role to restore French prestige in a world dominated by the "Anglo-Saxon" powers after World War II. De Gaulle encouraged the informal and then formal relations between the republic and Quebec. He was the first international statesman (unless one counts the flurry of interest shown by Napoleon III) to recognize Quebec for what it is, an island of French speech and culture in an Anglo-Saxon sea. His interest and that of the new nationalists of Quebec happily coincided. Moreover, de Gaulle had been rebuffed, largely because of U.S. influence, in two attempts to deal with Canada: first, in the attempt to sell French aircraft to Air Canada in 1964, and second, in an attempt to buy uranium from Canada in 1965 without giving assurance that it would not be used for military purposes. De Gaulle therefore saw in Quebec an occasion to raise French prestige in the world, and in Canada a chance to separate that country from what he regarded as American domination. He visited Quebec in the summer of 1967 and received an extraordinarily emotional reception. There, in an apparently calculated move, he took the opportunity to encourage Quebec *separatisme,* and in a speech in Montreal he created a furor by repeating the slogan of the French separatists: "Vive le Québec Libre!" De Gaulle was gently but directly rebuked on the spot by Mayor Jean Drapeau of Montreal, his official host, and firmly by the Canadian government, then under the administration of Lester B. Pearson, and he returned to France in customary grandeur. His *bêtise* (foolishness), however, was possible because there was a separatist movement of growing strength in Quebec.

Domestic developments. In many significant ways, World War II had been a watershed in Canadian history. The role of the federal government as engineer of national economic growth had been considerably strengthened. Yet there were complications and difficulties.

One problem was rooted in the expectations of Canadians for a better postwar world. When close to a million veterans reentered civilian life, they married, began to raise children (this was the start of the "baby boom" in Canada), and went on a buying binge. After all, Canadians had not really been able to enjoy themselves since the start of the Great Depression in 1929. That buying binge put tremendous pressure on Canada's balance of payments with the United States, because much of what Canadians were buying was built in the United States. It also added to inflationary pressures that stimulated industrial unrest, especially in 1945 and 1946. Organized labour had virtually doubled in size during the war; the unions were ready and willing to demonstrate their new strength, with major auto, steel, and transportation strikes.

King retired as prime minister and leader of the Liberal Party in 1948, and the mantle of leadership passed to Louis Stephen Saint Laurent, a Quebec lawyer whom King had brought into the government in 1941. Saint Laurent continued most of the domestic policies of his predecessor but was much more an activist in foreign policy. His time in office coincided with the intensification of the Cold War in the late 1940s, which brought an increase in defense spending. This, and opposition from provincial governments, eventually forced him and his government to curtail plans to expand existing social programs or to introduce such new ones as a national health insurance scheme. Saint Laurent was a popular leader, especially in Quebec, and was aided by a strong cabinet team and an effective civil service. He won major victories in the 1949 and 1953 federal elections, reinforcing the notion that the Liberals were somehow "the government party" destined to govern Canada forever.

Postwar prosperity. In the 20 years after 1950 Canada enjoyed growth and prosperity such as it had never known before. Not only did its old primary industries thrive, but Canada also embarked on a new phase of industrial development. In this, Canada was only moving with the greater industrial powers, especially the United States; the significance lay in it now being able to move in such company. The result was the development of electronic, aeronautic, nuclear, and chemical engineering on a great scale and the transformation of industrial Canada. Much of the growth, however, was also simple enlargement of earlier estab-

New mineral discoveries

lished industry, such as steel production. New sources of minerals were, indeed, part of the boom of the 1950s. Not only Labrador iron but also newly discovered deposits of radium, petroleum, and natural gas gave Canada resources it had theretofore had only in comparatively small supply. Mining investment also provided an excellent example of two matters of primary importance: first, how much the greater part of Canadian economic growth had come to be financed by American capital, largely in the form of direct investment and American ownership of the plant and control of operation, and, second, how foreign investment, largely American, aided by the American demand for Canadian materials, made the Canadian boom possible. Investment from abroad was, of course, eagerly sought, especially by the provincial governments, and Canada prospered both because of it and because of the resulting advanced technology and management. Nonetheless, Canadians were of two minds about U.S. investment. Many undoubtedly agreed with Saint Laurent's minister of trade and commerce, Clarence Decatur Howe, that the more U.S. investment in Canada the better. But others were uneasy over the growth of U.S. control over Canadian businesses and over the obvious partnership between Howe and American enterprises. Nowhere was this unease more apparent than in the House of Commons when, in May 1956, Howe tried to ram through a bill to finance a trans-Canada natural gas pipeline scheme, mostly backed by U.S. capital. The opposition created an uproar that politically weakened Howe and Saint Laurent.

Ethnic minorities. Much of the new economic development took place in Canada's northlands and had some part in ending the nomadic hunting life of the forest Indians and the Inuit of the Arctic shores and islands. This contact between the Canadian government and the Indians signaled a new dilemma that Canada faced in trying to deal equitably with its aboriginal peoples. After 1945 it was apparent that the old system for administering Indian affairs was not enough and that the system was itself collapsing. The poverty and disease of many reserves were a reproach to Canadian policy and humanity. The health care provided was greatly improved, and in 1959 the Indian Act was amended to increase opportunities for Indian influence on decisions affecting them. The métis, equal to whites according to the law, while in fact often treated as status-Indians, played an important part in the growing protest. The federal government took note by granting the franchise to all Indians in 1960, and several provinces followed suit.

Improved Indian relations

Large-scale immigration after World War II challenged Canada's social structure. First displaced persons, then people leaving war-devastated Europe, and then those coming to share the prosperity of Canada in the 1950s constituted the greatest immigration Canada had known. A majority of the immigrants were from Europe, many from southern Europe, particularly Italy, Greece, and Portugal; and almost all went into industrial cities in the east. As a whole the movement marked a new phase in the development and character of Canadian society. By 1961 the percentage of Canadians of British and Irish descent had fallen to 43.8 and of French descent to 30, while the percentage of persons of other descents had risen to 26.4. Moreover, most newcomers to Quebec chose to speak English and not French and to join the English culture. This stimulated the growing nationalism in Quebec.

Internal politics. The Pipeline Debate of 1956 revealed the malaise that was gripping the Liberal government after 21 uninterrupted years in power. Saint Laurent, though still personally popular, appeared to be old and tired. It was widely believed that he was losing his grip on the reins of government. Howe's actions during the Pipeline Debate, many felt, were an indication that he and other Liberal leaders had truly come to believe in their divine right to govern. The time was right for a new and dynamic Conservative leader, John George Diefenbaker, to end the decades of Liberal rule. A powerful orator, Diefenbaker challenged Canadians to open up the north, diversify their international trade, and end corrupt Liberal rule. In 1957 he was elected with a minority government, and the following year he won the largest parliamentary majority in Canadian history up to that time. During Diefenbaker's term of office, however, there was a major recession. He had to face the strains of an unsuccessful British attempt to enter the European Economic Community (Common Market); of relations with the United States during its Cuban missile crisis, in which Canada was not consulted and yet was expected to take part in the air defense of North America; and of a domestic struggle over the installation of nuclear warheads in Canada and their use by the Canadian NATO contingent. Internal dissension brought the Diefenbaker government down to a minority government in 1962 and to defeat in 1963. The Liberal government that followed was led by Lester B. Pearson, who had been chosen party leader following the 1957 election. Pearson included many talented men and women in his Cabinet, but many of them were also inexperienced. He initially set out to launch "one hundred days of decision," but he was stopped short when his finance minister, Walter Gordon, was forced to withdraw from his first budget controversial proposals to reduce U.S. investment in Canada. This and other blunders and scandals seemed to dog Pearson during his entire five years in office. Nevertheless, though he never achieved a majority government (he sought just that in a federal election in 1965), his government was one of the most productive in Canadian history. Under Pearson, Canada gained a national flag, a national social security system (the Canada Pension Plan), and a national health insurance program and granted federal public servants the right to free collective bargaining. While accomplishing all this, however, Pearson was also hampered by the rise of nationalism and separatism in Quebec, and he announced his retirement in 1967, to be succeeded in 1968 by Pierre Elliot Trudeau.

The Diefenbaker government

The origins of French separatism. The cause of the movement for a special status for Quebec within confederation, or a new form of association on the basis of equality with English Canada, or complete independence as a sovereign nation, was the belief, shared by a growing number of Quebec intellectuals and labour leaders, that the economic difficulties of Quebec were caused by confederation and could only be ended by altering, or ending, the ties with other provinces and the central government. In part it was the outcome of the perennial desire of Québécois to have their own state, as in a sense they had from 1791 to 1841. In part it was a continuation of the minority of French Canadian opinion that had opposed confederation in 1867. It was also a continuation of that sense of minority grievance stimulated by the execution of Riel, given substance by the Manitoba Schools Question, and given voice in the nationalism of Jules-Paul Tardivel and Henri Bourassa.

Origins of French nationalism

Yet French Canadian nationalism was also the outcome of profound economic and social changes that had taken place in Quebec since about 1890. Until that time French Canadians had lived by agriculture and seasonal work in the timber trade. The middle-class French of Quebec and Montreal acted as intermediary between their humbler countrymen and the English masters of commerce and industry. The coming of hydroelectric power and the woodpulp industry as a result of the successful national policy of protection in creating Canadian manufacturing plants in Quebec and Ontario created a labour force that brought French Canadians as workers into the cities, particularly Montreal. The rate of growth of the French Canadian population and the lack of good workable land outside the narrow St. Lawrence and Richelieu valleys contributed to the rush to low-paying jobs in urban industries and to the growth of slums, particularly in Montreal. By 1921 Quebec was the most urbanized and industrialized of all Canadian provinces, including Ontario, which remained, however, the most populous and the wealthiest. The Quebec government, devoted to the 19th-century policy of laissez-faire, recklessly encouraged industry and did little to check its worst excesses. With few exceptions the new enterprises were owned and directed, in the English language, by English Canadians or by U.S. businesses. It was a kind of second conquest of Quebec.

At the same time it destroyed the myths by which French Canada had survived: that of the Roman Catholic mission to the New World, and the cult of agriculture as the basis of virtuous life. Ever more Québécois had to make their way as best they could in quite a different world, in which the old values were mockeries and the obstacles not to be overturned by the old powers of endurance.

This clash of old and new came to a head in the last years of the regime of Premier Maurice Duplessis. A Conservative who became leader of the new Union Nationale Party formed during the Great Depression, he came to power in 1936, was defeated in 1939, but returned in 1944, and remained in office until his death in 1959. Duplessis, after 1944, wielded unequaled power in Quebec. Anglo-Canadians and U.S. big business furnished him the money he needed for political purposes; he in return preserved in the province a free enterprise system that would have done credit to a government before 1850. In the political miasma he created, all possibility of reform, all political integrity, seemed to be stifled.

Yet they were not. The fight against conscription in 1944 lifted Quebec politics out of the rut of place and favour. In 1949 the prolonged and bitter fight of the asbestos workers gave young idealists, one of them Pierre Elliott Trudeau, a chance to combine with labour in a struggle for a free society of balanced interests. Over the decade intellectuals began a searching analysis of Quebec politics and society in modern terms, often French and American as well as English Canadian.

The successors of Duplessis

When Duplessis died, it was a deliverance. Even his successors in office, Paul Sauvé and Antonio Barrette, seemed like reformers and to some degree were. But in the election of 1960 Jean Lesage and the Quebec Liberal Party made themselves the political vehicle of the new outlook and tried to carry out a Quiet Revolution—that is, to bring in much needed reforms before demands became revolutionary. They did much, both to assist the power of the province against the federal government and to transform Quebec society from clerical to secular and from 19th-century laissez-faire to positive, if not socialist, politics. Politically, indeed, they did too much for the taste of the rural areas of Quebec, and the party was defeated by the Union Nationale in 1966.

In the years that followed, the extremes widened in Quebec, and the elements of opinion began to crystallize. The Liberal Party was federalist; it held that the reforms needed in Quebec could be obtained within the federal system. So, at bottom, did the Union Nationale, but, as a mainly conservative party, it had to stress the importance of remaining Québécois and of obtaining greater provincial power. To the left of the traditional parties, however, opinion ranged from a demand for a special status for Quebec to the demand for separation and independence. The truth was that an active minority of Montrealers had passed far beyond the concept of constitutional reform, and in the process, of course, they broke with Trudeau and the Liberals. No social revolution (which, in their opinion, Quebec desperately needed) was possible, they believed, without independence. From their efforts came the Parti Québécois with a platform of secession from confederation. Under René Lévesque, a former Liberal, they won 24 percent of the popular vote in the election of 1970, but the Liberals, by reasons of the antiquated distribution of electoral districts, won 72 seats out of a total of 95. Constitutional reform was to be tried once more.

The growth of terrorism

Still other social revolutionaries, inspired by refugees from Algeria and by the Cuban example, began to practice terrorism. Bombings began in 1963 and continued sporadically. This mode of action most French and English Canadians felt was "un-Canadian," but it illustrated both the social ills of Quebec and the ties of the French intellectuals with the world outside Canada. In October 1970, with the kidnapping of James Cross, of the British consular service, and Pierre Laporte, minister of labour in Quebec, who was murdered, the government of Quebec called in the federal government for help, and the War Measures Act was proclaimed. The usual civil liberties were suspended, some 500 people were arrested, and troops were moved into Quebec. Few convictions followed—except those accused of the murder of Laporte—and public opinion in Canada, including Quebec, which in the majority approved the invocation of the act, began to wonder if the action had been too drastic.

The Trudeau years. *Domestic policies.* Pierre Elliot Trudeau, a member of Lester B. Pearson's Cabinet, moved in to fill the leadership void after Pearson's decline. He was a strong federalist who in 1968 won a decisive victory in Canada and Quebec. Whatever historians of Canada may decide as to the events and character of the years after 1968, it is certain that the political history of those years will be dominated by Trudeau. His government was highly personal. His ideas were clear, precise, and inflexible. His beliefs had the certainty of dogma. Never before had Canada been governed by a prime minister of personal assurance bordering on the arrogant and flavoured by the autocratic.

This influence of one man on the political life of Canada arose from two circumstances. One was the uncertainty introduced into Canadian politics by the rise of separatist feeling in Quebec. The other was the national feeling that Canada as a whole must take its own constitution and remake it to fit the circumstances of the late 20th century. Trudeau seemed superbly equipped to deal with both. He was a strong and convinced federalist, opposed flatly to separatism. He was a constitutional lawyer, acclaimed as an expert in constitutional understanding and reform. At the same time, he was impeccably French, the answer to the need of the Liberal Party for a French leader and to that of Canada for a French champion of the federal union.

Official Languages Act

As such Trudeau was free to complete Pearson's work in providing for a bilingual and bicultural Canada. To do this, his government introduced the Official Languages Bill in 1968 and issued the instructions under the resultant act, supported by all parties, which prepared the way for a bilingual federal civil service and for the encouragement of the French language and culture in Canada. (Similar encouragement was given to other ethnic cultures.) This was the foremost legislation of Trudeau's early years in office and was meant to begin a new relationship between the English and French in Canada.

Trudeau's government was also a government conducted, steered, and delicately attuned by the results of polls. Trudeau's personal aides ignored Canadian history and constitutional conventions; their eyes were fixed on the new political methods of the United States. Out of this personal and pollster government was to come a new constitutional concept, the doctrine of the personal constitutional mandate of the prime minister, a mandate measured by the life of Parliament, which had come only to reflect the image of a presidential prime minister.

Trudeau's special concern to maintain the unity of Canada and the good relations of English and French Canadians became the specialty of the Liberal Party in Quebec, both federal and provincial. Under a new leader, Robert Bourassa, the provincial Liberal Party in 1970 swept back into office. It was strongly committed to maintaining the federal system and to demonstrating the benefits of that system for Quebec. Compared with its electoral success and energetic policy of large investment and rapid economic development, the actions of the Parti Québécois seemed those of a petulant minority. The Bourassa government was at the same time strongly provincial and determined to make Quebec's viewpoint known and, if possible, prevail. Thus when the second of Trudeau's tasks taken over from Pearson, the reform of the constitution, had been, as it seemed, completed in the "Victoria charter" put before a meeting of the federal and provincial governments at Victoria, B.C., in 1971, Bourassa claimed a special position for Quebec. Despite agreement by all the other provinces, Quebec at the last moment withdrew its assent, and the path of constitutional reform was blocked.

At the same time a certain *folie de grandeur* led the province and the city of Montreal into enormous public expenditures. The chief result was that the Olympic Games of 1976 cost about $1.5 billion. Such recklessness augured ill for the federalist party of Quebec.

In grandiose expenditure, however, Quebec did no more than lead the way. The Trudeau government, following a policy of heavy public expenditure initiated by Pearson (and emulated by all the provinces), entered on a program of economic growth based in the first instance on government direction and expenditure. Government revenues (and, usually, deficits) rose to unprecedented proportions of the national income.

The French economy's coming of age

It was during the Trudeau years, however, and sometimes despite the policies of the Trudeau government, that the Canadian economy came of age. Canada continued to be a major supplier of foodstuffs and raw materials to the world during the boom decade of the 1970s, but Canadian secondary industry surpassed the primary industries as employers of people and exporters of products.

The west benefited greatly during the boom years. The products of mines and forests, on which the economy of British Columbia was so dependent, found ready markets at high prices in the United States and the Pacific Rim countries. One of the world's largest ocean coal depots, Roberts Bank, was built near Vancouver to expedite the shipment of British Columbian coal to Japan. Saskatchewan potash and uranium commanded premium prices during these years, and international demand for wheat, beef, and other farm products brought prosperity that matched the inflation-generated increase in land values. No province benefited more than Alberta, however, where escalating world oil prices brought wealth previously unimaginable and a tremendous land and construction boom—along with runaway inflation—in Edmonton and Calgary.

The increase in the price of oil and a rise in the price of natural gas sparked exploration in frontier areas such as the Beaufort Sea and the Arctic Archipelago. Some of this was aided by a variety of federal grant incentive programs and carried out by a joint federally and privately funded consortium, Panarctic Oils Ltd. Fearing that foreign capital would permanently dominate the Canadian oil industry, the Trudeau government created the integrated, crown-owned Petro-Canada in 1975.

In central Canada prosperity kept pace. The Canada–United States Automotive Products Agreement (Autopact), concluded in 1965, finally began to pay dividends as U.S.-owned carmakers began to build new assembly plants in Ontario and Quebec. Soon tens of thousands of new jobs were created in the automobile and the auto parts industry, and Toronto quickly passed Montreal as the financial capital of Canada. Although much of Atlantic Canada was still mired in poverty, it too experienced a taste of prosperity as foreign auto and tire plants began to locate there to take advantage of Autopact.

All of this growth came at a price: inflation. By the mid-1970s the Trudeau government was becoming preoccupied with the fight against rising prices and the wage increases that usually followed. In this it was in step with most of the Western industrialized countries. In 1975 the federal government created the Anti-Inflation Board and imposed wage and price controls for a three-year period. The move had the general support of business but incensed the labour movement, which called for a day of protest—actually a national general strike—in October 1976.

Foreign affairs. In 1970 the Trudeau government issued a major statement defining its foreign policy. Three major aims were presented: preservation of Canada as an independent political entity, maintenance of expanding prosperity, and constructive contribution to human needs.

Relations with China

In 1970–72 Canada reduced the number of its military and civilian personnel and military bases in Europe while remaining within NATO. Canada established diplomatic relations with the People's Republic of China on Oct. 13, 1970; ambassadors were exchanged the following year, and an exchange of consuls and most-favoured-nation trading arrangements were agreed upon in 1973. Canada also sought closer relations with the Soviet Union and the European Economic Community and played a more active role in the United Nations. Canada maintained its commitment to the North American Air Defense Command (Norad), but in 1975 regional responsibility was reorganized to include only Canadian territory within Canada's jurisdiction. In 1981 the Norad agreement was renewed, and its name was changed to North American Aerospace Defense Command.

Protection of Canada's economy led to adjustments in relations with the United States. In 1970 Canada increased the price of petroleum and natural gas sold to the United States, and in 1974 a plan to gradually reduce those sales and end them by 1982 was announced. This action was taken to protect domestic supplies of fossil fuels in the face of increasing prices of imported oil used in the eastern provinces. In 1978 Canada initiated purchases of new airplanes and other military equipment to better defend its borders and fulfill its international commitments.

In accordance with the third aim of its foreign policy—to contribute to human needs—Trudeau's government pursued a policy promoting the international control of nuclear weaponry. In 1972 Canada and the United States signed the Great Lakes Water Quality Agreement to control pollution of the lakes.

In and out of power. By the late 1970s the glamour of the Trudeau regime was wearing off, and the policies his government had pursued were falling into confusion. The bilingual policy was found to have been pushed beyond the brink of tolerance in English Canada and was hastily truncated before even the remodeling of the federal civil service was completed. The assertion of a strong central federal government was shocked by the sudden electoral victory of the Parti Québécois in Quebec provincial elections in November 1976. The grounds of its victory were the corruption and mismanagement of the Bourassa government, but the party was committed to separation, at least in the form known as sovereignty-association. Attempts to amend the constitution had failed, apparently beyond hope of recovery.

Growth of provincial government

The reasons for this slow but powerful reversal of much that a brilliant prime minister had stood for could be found in the enormous expansion of skill and power in provincial governments. There had been a steady widening of provincial jurisdiction (especially in the field of welfare), an expansion of revenues and expenditures, and a growing sense of local importance partly nurtured in the constitutional device of the federal-provincial conference. The provinces for the first time in 1975 spent more of the national product than the federal government did. The federal government had now become less powerful than the provinces acting collectively. These were more and more inclined to act collectively as more and more of them came to have Conservative or opposition governments—9 out of 10 by the fall of 1978.

Although Trudeau had won a solid majority in the 1968 federal election, much of his power base in western Canada, Ontario, and the Atlantic Provinces began to dissolve soon after. His almost constant preoccupation with Quebec and Quebec-related issues, combined with his seeming lack of sympathy with regional concerns, eroded his popularity. He was reelected only with a minority in 1972 and was forced to rely on the social democratic New Democratic Party to govern. Although he was able to refashion a majority in 1974, his victory was as much a result of dissatisfaction with the Conservative opposition as it was an indication of his popularity. In fact, the next five years witnessed a drastic decline in Liberal popularity. Trudeau's wage and price controls alienated organized labour. His advocacy of greater government intervention in the economy alienated business. His constant efforts to impose Ottawa's will on the provincial governments disturbed voters in Atlantic Canada and the west. He waited as long as he could, but in May 1979 the required election was held, and Conservative leader Joe Clark took power with a minority government.

Clark was the youngest Canadian prime minister ever—he was not quite 40 years old when elected. His youthful inexperience showed all too plainly in the months that followed. One of his major policy blunders turned out to be a promise made during the election to move Canada's embassy in Israel from Tel Aviv to Jerusalem. When he undertook to keep that promise after the election, an uproar led by federal bureaucrats and business leaders who feared an Arab boycott of Canada forced him to reverse himself. Perhaps his most serious mistake, however, was his attempt to increase the federal gasoline tax as a means of increasing Ottawa's share of the windfall oil profits that had flowed from the rise in energy prices and of promoting conservation of gasoline. When Clark's budget containing the tax was voted down in December 1979, his government was defeated. In February 1980 Trudeau was reelected with a majority, even though he had contemplated stepping down as Liberal leader after his defeat the previous May. Trudeau quickly consolidated his power and took up where he had left off. His continued opposition to separatism was evident when he took an active part in the May 1980 referendum in Quebec called by the Parti Québécois government in an effort to secure a provincial mandate to negotiate sovereignty-association—a watered-down form of independence—with the rest of Canada. That intervention helped tilt the balance against the pro-separatism forces.

The Canada Act

Following the referendum, the Trudeau government renewed its efforts to secure for Canada possession of and authority over its own constitution. The issue centred on the Constitution Act of 1867, which could be amended only by the British Parliament on Canada's behalf. The issue was complicated by the need to adopt an amending process acceptable to the federal government as well as to the 10 provinces. On Dec. 2, 1981, an amending process and a bill of rights (Charter of Rights and Freedoms) were accepted by all except Quebec. Nevertheless, the British Parliament approved the resolution on March 25, 1982; and Queen Elizabeth II on April 17, 1982, issued a proclamation making Canada fully independent and designating Canada's constitutional document the Canada Act. Canada retained its status as a member of the Commonwealth. Quebec, however, would not accept the new document without wording that would give that province a special status on the basis of its French heritage.

The patriation of the constitution and the adoption of a charter of rights was a political triumph for Trudeau and the culmination of a career-long campaign to put civil rights and liberties in Canada above the reach of legislatures. Trudeau's economic performance in his last years in power was somewhat less of a success. Canada suffered greatly in the worldwide recession of 1981–82, but the impact was made worse by Ottawa's failure to control its spending and its miscalculation in anticipating that future increases in energy prices would help pay its bills. That expectation was the basis of the National Energy Program (NEP), introduced in the fall of 1980, which was designed to speed up the "Canadianization" of the energy industry and vastly increase Ottawa's share of energy revenues. The NEP sparked a war with the energy-producing provinces, especially Alberta, chased private investment capital out of Canada, and drastically reduced exploration for conventional oil and gas. When oil prices declined, NEP policies made the recession even deeper in Alberta than virtually anywhere else in Canada.

In 1984 Trudeau resigned and was succeeded as head of the Liberal Party and as prime minister by John Turner. Turner immediately called for elections to be held in September; at that time the Progressive Conservative Party gained a parliamentary majority in a landslide victory. Party leader Brian Mulroney became prime minister. Mulroney's approach to government was, in most respects, very different from that of Trudeau. In federal–provincial relations he sought to avoid the bitterness and rancour that had marked Trudeau's dealings with the provincial premiers. Accords were negotiated with Newfoundland and Alberta that ended the crisis over federal energy policy and dismantled the NEP. In November 1984 the finance minister, Michael Wilson, announced that the government would adopt a new approach to economic and fiscal matters to encourage private, including foreign, investment, to bring down the national debt, to review social programs, and to sell Crown corporations (*i.e.,* privatization). It was in this last area that the most progress was made.

The government's pledge to reduce the national debt was not fulfilled. No major social programs were ended, and, although some savings in government expenditures were put into effect, the deficit was in fact controlled through major tax increases.

The Meech Lake accord

Two major initiatives marked the government's first period in office: a constitutional agreement with all 10 provinces known as the Meech Lake accord, designed to bring Quebec's agreement to the Canada Act of 1982; and the Canada–U.S. free-trade agreement. The Meech Lake accord was concluded in the spring of 1987, but response for ratification from some of the provinces was slow. Newfoundland's and Manitoba's refusal to ratify the accord by the June 23, 1990, deadline was a severe blow to Mulroney and created a new crisis on the issue of Quebec separatism.

Mulroney was more successful with the free trade agreement. Negotiated with the United States over a period of two years, the agreement was signed by Mulroney and U.S. President Ronald Reagan in January 1988. It easily passed through the U.S. Congress but was the object of bitter debate in Canada. In the federal general election held in the fall of 1988, Liberal opposition leader John Turner accused Mulroney of selling out Canada's birthright and mounted a major assault on the agreement. Free trade was virtually the only issue in that election, which Mulroney won with a reduced mandate but with his majority intact. On Jan. 1, 1989, the free trade agreement went into effect. (W.L.M./D.J.Be.)

For later developments in the history of Canada, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 964, 965, and 973, and the *Index.*

ATLANTIC PROVINCES

New Brunswick

A Canadian province located on the eastern seaboard of the North American continent, New Brunswick is Canada's most forested province and also its only officially bilingual one, French and English having equal status. It was one of the four original provinces making up the national confederation in 1867. Together with Nova Scotia and Prince Edward Island, it forms the regional grouping known as the Maritime Provinces. While New Brunswick has contributed heavily to the national life of Canada, it is one of the smaller provinces and has always occupied a lesser role in the national economy. Its beautiful forests, rivers, lakes, and seashore—attractions for tourists, hunters, and sports fishermen—have remained relatively unspoiled. The province has a roughly rectangular shape, about 210 miles (340 kilometres) from north to south and 185 miles (300 kilometres) from east to west, and it has an area of 28,355 square miles (73,440 square kilometres). It is bounded on the west by the U.S. state of Maine, on the north by the province of Quebec, on the east by the Gulf of St. Lawrence and the Northumberland Strait, and on the south by the Bay of Fundy. The latter water bodies are separated by the narrow neck of the Chignecto Isthmus, which joins New Brunswick to Nova Scotia. Chartered by King George III, the province was named after the royal House of Brunswick, and its capital, Fredericton, for King George III's son.

PHYSICAL AND HUMAN GEOGRAPHY

The river system

The land. The land rises abruptly and in some places spectacularly out of the Bay of Fundy in the south and undulates northward in a series of rolling hills and rocky outcroppings. Plains and flatlands are limited to sea-level marshlands in the southeast, river valley floodplains, and low-lying lands adjacent to the beaches of the east coast. There are no true mountains, although ranges of rugged and sometimes steep hills run east–west a few miles above the Fundy coast and in the north-central part of the province, where the highest elevation is Mount Carleton at 2,690 feet (820 metres). This hilly landmass is cut in every direction by an extensive river system. The St. Croix follows the southwestern boundary. The 418-mile- (673-kilometre-) long St. John, with its numerous tributaries, drains the entire northwestern, central, and south-central parts of the province. The high tides of the Bay of Fundy reverse the rapids at the river's mouth at Saint John and raise the level of the lower river for up to 95 miles (150 kilometres) inland. Additionally, a number of lake-size arms branch off the lower St. John on its eastern side, the largest of which is Grand Lake, about 60 miles (100 kilometres) inland. The eastern part of the province is drained by the Restigouche River in the north, the many branches of the Miramichi in the north-central area, and the shorter Petitcodiac in the southeast. Spring flooding has always been common in many of the rivers. The majority of soils are acidic and low in nutrients. Most of the well-drained soils would be classified as podzols or luvisols. Almost all cultivable soils have moderate to severe limitations that restrict the range of crops or require special conservation practices.

Climate. Clearly distinguishable seasons characterize the climate. Winters are snowy and cold, and summers mild and pleasant. The fall is particularly spectacular, with alternating cold nights and sunny warm days turning deciduous foliage into a blaze of colour contrasting against the solid dark greens of the conifers. Temperatures at Fredericton range from an extreme low of −35° F (−37° C) to an extreme high of 99° F (37° C). In January the mean daily temperature is 15° F (−9.2° C), and in July 67° F (19.3° C). Temperatures are more moderate in the southern coastal regions. Normal annual rainfall is 32 inches (825 millimetres), and snowfall averages about 114 inches (2,900 millimetres); annual precipitation amounts to about 43 inches (1,100 millimetres).

E. Otto/Miller Comstock Inc.

Forested hills rising behind a small farm near Waterford, N.B.

Plant and animal life. Forests cover about 90 percent of the province, and coniferous trees commonly found in north temperate climates—such as balsam fir, red and black spruce, pine, and hemlock—constitute the principal growth. Deciduous stands, especially along the river valleys, consist of sugar and red maple, trembling aspen, and yellow and white birch. Wild blueberries and cranberries are found commonly on thin soils, especially in southwestern areas. The forests support substantial herds of white-tailed deer, moose, and small animals, such as porcupines and raccoons. A regulated hunting season occurs in the fall of the year. Several rivers are famous for their Atlantic salmon, while trout, bass, and pickerel also attract sport fishing. Whale watching near the western islands of the Bay of Fundy has become popular. The rivers and coasts lie in the eastern flyway of numerous migrating birds, best seen in early to mid-spring and late summer.

Towns

Settlement patterns. Settlement began on the coastline and rivers, where early settlers followed fishing and agricultural pursuits. Small towns tended to spring up along the river systems, which supported sawmills and general stores. Saint John, an all-season ocean port situated at the mouth of the St. John River, has been the largest city from earliest times. Moncton in the southeast emerged in the late 19th century as a railroad and distribution centre. Fredericton, the capital and an administrative, educational, and commercial city, is near the geographic centre of the province. Bathurst, Campbellton, and Edmundston in the north are small bilingual cities with French-speaking majorities, all developed on the basis of the pulp and paper industry. In the 1950s the Canadian Armed Forces established Base Gagetown, covering several thousand acres between Fredericton and Saint John, and they built the modern town of Oromocto for army personnel.

The people. The English-speaking majority consists largely of the descendants of American loyalist, Scottish, Irish, and English settlers of the 18th and 19th centuries, and it is concentrated in the southern and western parts of the province. The French-speaking minority, which has grown to about one-third of the population, consists of the descendants of 17th-century Acadian settlers augmented by French Canadians from Quebec, and it is concentrated in the northern and eastern counties. In addition, the province's first inhabitants, the native Indians, remain in small numbers on federally administered reserves along the east coast or in the St. John valley, or they have been assimilated into neighbouring communities. A handful of blacks are largely descendants of loyalist slaves. In the 20th century, small contingents of eastern European, Dutch, German, Italian, and South and East Asian immigrants added a multicultural dimension, mostly in the larger urban areas. Roman Catholics outnumber Protestants, who are mostly of the Baptist, United, Anglican, Pentecostal, and Presbyterian denominations.

Out-migration

Population growth was relatively slight in the late 20th century. A steady out-migration of young adults, drawn by the employment opportunities and higher salaries of central Canadian cities, continued to characterize the society. Within the province, cities became the employment centres, although the common practice of commuting from rural homes or from suburban areas prevented significant urban population growth.

The economy. New Brunswick has a resource-based economy based largely on forestry, mining, and fishing. Tourism, agriculture, small-scale manufacturing, and a growing service sector provide balance and diversity. Employment levels, however, are subject to seasonal variation, and the economy remains dependent on federal government programs. Such programs support the provision of economic infrastructure such as in transportation and communication and have contributed substantially to the development of up-to-date amenities and services comparable to those found elsewhere in Canada. Real economic growth tends to be above the regional average, reflecting expansion in mining and electrical power exports.

Agriculture. Only 20 percent of the landmass of the province is suitable for agriculture, and less than 6 percent is actually under cultivation. Small family farms remain the most numerous type among the fewer than 3,600 farms, but their numbers are dwindling and thousands of acres of previously cultivated land are returning to forest. The largest farms specialize in the leading cash crops, dairy products and potatoes, the former concentrated in south-central and southeastern areas and the latter in the upper St. John River valley. Both benefit from large-scale food processing plants in their respective areas. Poultry, cattle, hogs, eggs, maple products, and blueberries represent other significant items. Truck farming is concentrated near the southern cities, such as in the lower St. John valley, where floodplains possess the richest soils. They represent about 2 percent of the provincial landmass. Provincial policy focuses on soil conservation and improved drainage to increase production.

Forestry and fishing. The forest industry is the province's largest, and pulp and paper production represents its largest single component. Several cities and towns, mostly in the north, depend on large pulp and paper mills as their major employers. Sawmills are widespread. The forests supply the material for lumber, plywood, chipboard, fuel, and Christmas trees and provide a wide range of jobs in related fields.

North Atlantic fishing areas

Fishing and fish processing continue to be important in the provincial economy, employing numerous workers in the main fishing areas, including the Bay of Fundy, the Northumberland Strait, and the Gulf of St. Lawrence. Lobster, crab, scallops, herring, cod, mackerel, and a variety of other North Atlantic species make up the bulk of the catch. The industry is a volatile one, subject to seasonal and longer-term shifts in fish stocks, foreign competition, weather conditions, and credit availability; it has thus become highly dependent on government regulation and support. In the late 20th century the development of aquaculture expanded the fisheries with hatcheries producing salmon, trout, mussels, and oysters.

Mining and energy. The discovery in New Brunswick of rich deposits of zinc and potash in the second half of the 20th century elevated the importance of mining in the provincial economy. The northeast, especially around Bathurst, produces zinc, copper, lead, silver, and peat moss. A lead-smelting operation is at Belledune, a port near Bathurst, and potash is found near Sussex. Nationally, the province ranks high in the production of antimony, bismuth, potash, peat, zinc, silver, and lead. Coal, gypsum, and sulfur are also significant.

Industrial development was spurred by the vigorous expansion of electrical power generating capacity by the New Brunswick Electric Power Commission, the provincially owned utility. Hydroelectric developments along the St. John River system were supplemented by coal- and oil-fired steam generating units and by the construction of a nuclear power station at Point Lepreau southwest of Saint John. A power grid links the New Brunswick utility with all the neighbouring provinces and states and permits the export of excess capacity.

Transportation. At Saint John, a modern ocean port and the province's largest industrial centre, oil refining, shipbuilding, and papermaking are major employers. The city's port facilities are able to handle container ships, and they play an important role in exporting Canadian goods, especially in winter months when traffic on the St. Lawrence Seaway is curtailed. Moncton and Saint John remain major transportation and communication centres despite the reduced importance of railroads, and Moncton has emerged as an important regional distribution centre. Regional and national airlines provide regular service to all the larger cities, and a modern bus service links most urban centres. Highway traffic tends to be heavy on routes linking the other Maritime Provinces with Central Canada and the United States.

Administration and social conditions. The provincial government follows the constitutional forms of a Canadian province within the framework of the Canadian constitution. The one-chamber legislature at Fredericton has 58 members, while the executive consists of the premier and a cabinet of about 20 members. The three national political parties are active, but only Liberals and Progressive Conservatives have formed governments; the New Democratic Party established a small foothold in the late 20th century.

Education

Public education is free and includes 12 grades of elementary and high school. The system, unique in Canada, is administered by separate English and French sections of the provincial Department of Education and is assisted by locally elected school boards with limited authority. Teachers negotiate their salaries with the province, which raises and distributes funds to school districts on the basis of equality so that all areas of the province enjoy similar standards. All universities receive public funding. The University of New Brunswick (established in 1785, one of Canada's oldest) operates from its main campus in Fredericton, which it shares with the Roman Catholic St. Thomas University; a second campus is in Saint John. The University of Moncton is French-speaking. Mount Allison University, privately funded in part, is in Sackville. There are several trade and technical schools.

The province has a network of modern hospitals, the largest of which are regional hospitals with specialized facilities and equipment. Hospital and medical care are both publicly funded and are supported by a provincial sales tax. Social services are administered by the provincial government, and the same standards apply provincewide. A provincial housing authority plans and develops projects for low-income families and for the elderly. Health, education, and social welfare programs are all heavily dependent on federal financial support.

Cultural life. Culture is shaped by the predominance of the two largest linguistic groups, English and French. Ethnicity remains a strong determinant of cultural activity in some areas, as in the annual Irish festival held at Newcastle near the Miramichi River, the blessing of the Acadian fishing fleet at Caraquet, and Loyalist Days in Saint John. Summer festivals and fall fairs retain the flavour of the past and the traditions of the province's early settlers. Popular tastes in music include Irish, Scottish, and Acadian folk idioms and their North American "country" derivatives.

Fine arts

The fine arts are distinguished by some outstanding work in painting, choral singing, and creative literature. Saint John artists of the 1930s and '40s, such as Miller Brittain and Jack Humphrey, produced striking social commentaries in their depictions of depression and war, while painters at Mount Allison University led the way in the 1960s in the development of a widely recognized regional group of magic realists. Regional and provincial music festivals encourage music performance among youth. Fredericton maintains its claim to be the poet's corner of Canada with the publication of the literary magazine *The Fiddlehead* and the work of Fred Cogswell and others, carrying on a tradition first established in the late 19th and early 20th centuries by Bliss Carman and Sir Charles Roberts, arguably the founders of Canada's first school of poetry. Antonine Maillet, Acadian novelist and playwright from Bouctouche, achieved international recognition for her writing in French, which strikingly reveals the 17th-century idiom and structure of the language as spoken by today's Acadians.

Several institutions provide leadership in the arts. Theatre New Brunswick and the University of Moncton promote drama. The Beaverbrook Art Gallery in Fredericton houses an excellent small collection of British paintings of the past three centuries, and in several towns there are other public and private galleries featuring mostly regional work. The New Brunswick Museum in Saint John, the oldest in Canada, contains archival treasures from the province's past, while two historic reconstructions—King's Landing Historical Settlement near Fredericton and Acadian Village near Caraquet—depict rural life of the past century. Such historic sites as Fort Beauséjour National Historic Park near Sackville remind visitors of the Indian, French, and British roots of the province. A provincial agency fosters cultural activity. Scholars rely on university libraries, the archives of the New Brunswick Museum, the Provincial Archives in Fredericton, and the centre for Acadian studies at the University of Moncton. Municipal libraries are widespread.

Recreation

Recreational activities fall into two groups, winter and summer, with greater emphasis on the former. Skiing, hockey, and curling dominate in winter. Summer recreation focuses naturally on water: Boating, fishing, and swimming flourish in a province that abounds with lakes, rivers, and accessible seashore, including miles of clean sandy beaches along the east coast. Golf, tennis, and hunting are also popular, while two national parks—Fundy on the rugged coast of the Bay of Fundy (featuring the world's highest tides) and Kouchibouguac on the Northumberland Strait—and several provincial parks provide camping, bird-watching, hiking, and other outdoor pursuits.

Several major daily newspapers circulate in the province. Among the leading English-language dailies are the *Telegraph-Journal* and *Evening Times-Globe* (Saint John), the *Gleaner* (Fredericton), and the *Times-Transcript* (Moncton). The main French-language daily is *L'Acadie Nouvelle* (Caraquet). Television broadcasting derives from three main sources, the Canadian Broadcasting Corporation (CBC), CTV Television Network, and Radio-Québec. Cable and pay TV are also available. French- and English-language radio stations broadcast throughout the province.

HISTORY

The French were the first Europeans to lay claim to the province, part of a larger region that they called Acadia (French: Acadie), which was inhabited by Algonquian-speaking Indians of the Micmac, Malecite (Maliseet), and Passamaquoddy tribes. The British took over in 1713, expelled or dispersed most of the French-speaking Acadian settlers in 1755 (many of whom eventually returned), and governed the area as Nova Scotia until 1784, when New Brunswick was established as a separate province with its present boundaries.

St. John River valley settlement

The first English-speaking settlers, from New England, moved into the St. John River valley and founded the town of Maugerville in 1762. But it was the influx of some 14,000 loyalist refugees from the American Revolution, mostly from New York and its vicinity, that created the pressure for separate provincehood. The loyalist city of Saint John became Canada's first incorporated city in 1785, and smaller settlements were established in the St. John and St. Croix valleys.

After early problems of adjustment the loyalist communities of New Brunswick began to prosper. Underlying the improved economy was the British decision in 1808 to grant preferential tariffs to the timber resources of its North American colonies, a move made when Napoleon's blockade cut off the Baltic supply of shipbuilding materials to the British. For New Brunswick, with its limited agricultural lands but widespread forests, this historical incident provided an opportunity that helped usher in the so-called age of wood, wind, and water, an era of prosperity based on timber exports and shipbuilding. A Reciprocity Treaty with the United States in 1854 and the demand created by the U.S. Civil War further stimulated trade.

Politically, the province moved slowly toward a more democratic system. Government in the first generation was dominated by a loyalist elite supported by British imperial authority. Responsible home government was granted by Britain in 1848. Political parties emerged in the 1850s, organized largely along ethnic lines and prompted by the prohibition issue. With its economic success and political independence, New Brunswick entered confederation with Canada in 1867 only reluctantly.

Confederation, however, coincided with the collapse of the age of wooden ships, and New Brunswickers found themselves scrambling over the next 100 years to rebuild their economy on a new foundation. Railroads and national tariffs helped in the development of manufacturing, such as in cotton textiles, but the province suffered from the pull of the urban growth areas of Quebec and Ontario. By the 20th century, New Brunswick required federally subsidized freight rates, and it argued, along with other Maritime Provinces, for federal financial assistance, which, after the mid-1900s, it obtained.

Liberal program of the 1960s

Meanwhile, a modern party system emerged in which, until the 1970s, the old elements of ethnicity and religion continued to be significant. Liberals and Progressive Conservatives alternated in government at fairly regular intervals. In the 1960s, a Liberal equal opportunity program revolutionized the delivery of health, justice, education, and social services by abolishing counties as administrative units and by centralizing funding and administration at the provincial level. Unique in Canada, the system met strong early resistance, but it continued under the subsequent Conservative government and won acceptance, especially in less prosperous rural and northern areas, as a means of equalizing services in all parts of the province.

New Brunswickers think of their province, with its two languages and cultures, as a microcosm of Canada. Moreover, the province's small size and its reliance on federal government transfers tend to promote support for nationalistic approaches on constitutional and political matters. Even in the late 20th century, the province retained qualities of its rural and small-town past.

(W.S.MacNu./S.E.P.)

Newfoundland

The most easterly part of North America, the Canadian province of Newfoundland is composed of the island of Newfoundland and a larger mainland sector, Labrador. It is Canada's newest province, having joined the confederation only in 1949. The island, named the New Found Land by 15th-century explorers, has an area of 43,548 square miles (112,790 square kilometres) and lies athwart the Gulf of St. Lawrence. It is separated from Labrador by the narrow Strait of Belle Isle and from Nova Scotia by Cabot Strait. Labrador, with an area of 113,101 square miles (292,930 square kilometres), is bordered to the west and south by the province of Quebec. Newfoundland's position on the Atlantic gives it strategic importance in transportation and communications. Its capital city, St. John's, for instance, is closer to the coast of Ireland than it is to Winnipeg. Of perhaps greater significance are the great fish stocks inhabiting the Grand Banks to the east and south of the province, which spurred the development of numerous communities stretched along 10,000 miles (16,000 kilometres) of deeply indented, wave-battered seacoast. These fisheries have been the single most important factor in shaping the history and character of the land and its people.

PHYSICAL AND HUMAN GEOGRAPHY

Geologic characteristics

The land. *Relief.* Newfoundland and Labrador must be treated as two separate physiographic regions. The island, roughly triangular in shape, is part of the Appalachian geologic province of North America, in which the landforms run from southwest to northeast and are characterized by continental drift, volcanic action, crustal deformation, ice erosion, and deposition. These forces have produced a highly complex geologic structure with Precambrian rocks of ancient Europe and Africa on the east, newer Appalachian rocks on the west, and the bed of the ancient ocean squeezed up between them. On the west coast the land rises abruptly from a narrow coastal plain to a plateau with a maximum height of 2,651 feet (808 metres) that slopes away to the northeastern coast with its many headlands, islands, and bays. The plateau is gently undulating and dotted with thousands of lakes and ponds, numerous streams, and rivers, including the Exploits, Gander, and Humber. The coastal terrain is hilly and rugged. The southern coastal areas have numerous bays, fjords, and offshore islands.

Labrador is geologically part of the Canadian Shield, which comprises some of the world's oldest rocks. Although most of the rocks are Precambrian igneous and metamorphic formations, the Labrador trough, in the west, contains softer sedimentary deposits and includes some of North America's most extensive iron ore deposits. In the far north the Torngat Mountains rise abruptly from the sea to a height of 5,320 feet. The interior is like a giant saucer dotted with lakes and dissected by rivers that break through the eastern saucer rim to discharge into the Labrador Sea. The indented coastline has countless offshore islands, fjords, and coves; exposed and barren headlands; and relatively lush river valleys.

Most of the province shows the effects of continental glaciation during the Pleistocene Epoch. Thus, the material that lies underneath the thin layer of today's soil is generally glacial debris or marine sediments exposed by postglacial uplift. Nevertheless, the watersheds of the larger rivers contain deeper surface deposits that support the growth of excellent forest stands, and the province has pockets of arable mineral soils that support agriculture. Interrupted drainage systems throughout Newfoundland island have created extensive peat bogs.

Temperatures

Climate. In general, Newfoundland has cold but not severe winters and warm to cool summers. The mean July temperature of the coast ranges from 40° F (4.4° C) in northern Labrador to 59° F (15° C) on the island's south coast. In the southern interior the July mean is just above 60° F (16° C). January means in the southern portion of the island are above 20° F (−6.7° C), on the coast of Labrador approximately 10° F (−12.2° C) and 0° F (−17.8° C) in the south and north, respectively, and in interior western Labrador about −15° F (−26° C). Extreme low temperatures in southern Newfoundland rarely reach 0° F, but in western Labrador readings below −40° F (−40° C) are not uncommon. The annual precipitation varies from 55 inches (1,397 millimetres) in the southern parts of the island to about 17 inches (432 millimetres) at Cape Chidley. In the northern regions, as much as half of the annual precipitation occurs as snow. In the south, the snowfall usually accounts for only 20 percent of the total precipitation.

Mid-latitude storms, moving across Canada and up the Atlantic seaboard, have a pronounced effect on Newfoundland's climate. Warm air drawn in on the southern side of the disturbances contributes to the high precipitation in the southern part of the island. Northeasterly and easterly winds prevailing in advance of each storm blow across the cold Labrador Current, keep coastal summers cool, and impede the coming of spring. Northwesterly winds, following in the wake of each storm, carry in cold Arctic air that forces temperatures even lower than the latitude and marine location would suggest. To the south the cool air of the Labrador Current mixes with the warmer air over the Gulf Stream to produce frequent dense fogs, which are most common over the Grand Banks and along the south and southeast coasts.

Plant and animal life. With the exception of the tundra of northern Labrador and the barren reaches of higher altitudes and of some coastal regions, much of the province is forested. The principal species are conifers, of which balsam fir and black spruce are the most abundant. In most parts of the province, but particularly on the island, conifers are mixed with such deciduous species as the paper and yellow birches and a wide variety of hardwood shrubs. The best stands of forest occur in areas of deep and well-drained soils, while in less-favoured areas much of the forest growth is stunted. In others, repeated fires and subsequent erosion have created barrens that produce a great variety of small woody plants.

The moose is the most plentiful of the large wild animals found on the island, where they outnumber the herds of woodland caribou. In Labrador the case is reversed. Both, however, are outnumbered by herds of harp and hooded seals, who migrate along the coasts of the province. Other species include the black and polar bears, Arctic and red foxes, beaver, lynx, and the range of small fur-bearing animals common to the northern coniferous forests and the tundra of northern Labrador. Whales, now protected, are commonly seen throughout the summer as they feed and disport themselves in coastal waters.

Birdlife

Vast colonies of seabirds, notably murres, Atlantic puffins, northern gannets, petrels, and eider ducks, inhabit the offshore islands and headlands. Several species of gulls and terns are ubiquitous, and substantial breeding populations of black ducks and Canada geese are maintained, together with lesser populations of other ducks. Migratory shore and wading birds frequent the coast seasonally. Upland game birds include ptarmigan, grouse, and snipe, while such birds of prey as the osprey and bald eagle are common.

Settlement patterns. *The Avalon Peninsula.* Newfoundlanders see their province as comprising nine regions, distinct in their natural settings and human imprint. The most densely populated part of the island, the Avalon Peninsula on the southeast coast, contains the capital city, St. John's, the adjacent city of Mount Pearl, a number of smaller towns such as Conception Bay South, Harbour Grace, Carbonear, Trepassey, and Placentia, and a host of small villages typically located at the edge of the sea. These villages offer suitable harbours and such essentials as wood and fresh water, but most particularly afford ready access to near-shore fishing grounds. The neat and brightly painted houses, surrounded by small gardens and meadows, are generally of frame and timber construction, as are the adjacent stores, stages, wharves, and slipways. The helter-skelter arrangement of houses and other domestic buildings and of gardens, meadows, pastures, and woodlands suggests a delightfully picturesque and romantic nonconformity. Only after about 1950 was the idea of town planning introduced.

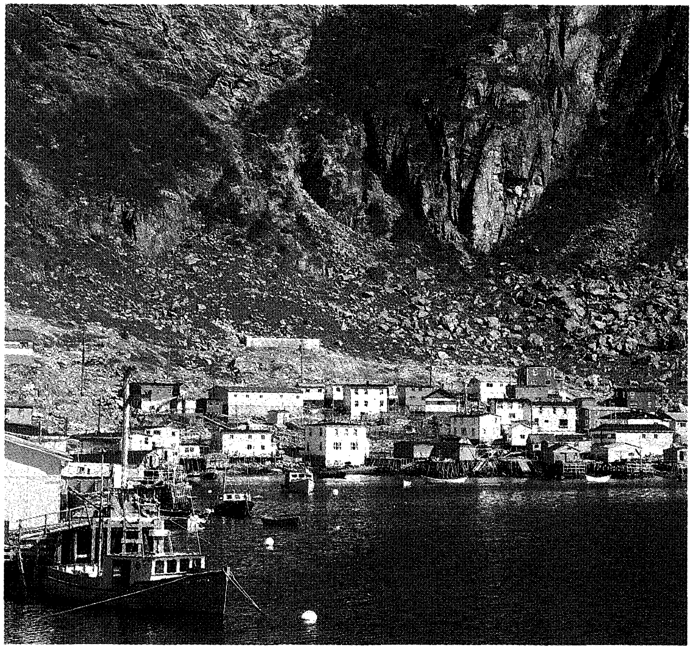
The east and northeast coast. Northward from the Avalon, along the coast rimming Trinity, Bonavista, and Notre Dame bays, the landscape is softened. The headlands and islands are still barren, but in the bays and estuaries the size and variety of the forest cover attest to deeper soils. The rocky shoreline is interrupted by stretches of sandy beach, and the fishing village occasionally is replaced by a farming community. Many islands shield the mainland from the worst effects of the northeasterly winds and create hundreds of miles of virtually landlocked waters. Many old communities such as Fogo, Bonavista, and Twillingate developed in the more exposed locations, where the exigencies of the fishery demanded that they be built. These are being outstripped in growth by the newer towns located in more favourable settings, such as Clarenville, Lewisporte, and Springdale, which have become important distribution centres.

Long Range Mountains

The Great Northern Peninsula. The Great Northern Peninsula stretches northward toward the Labrador coast. On the west, the Long Range Mountains, rising abruptly from a narrow coastal plain dissected by numerous fjords and rivers, offers spectacular scenic attractions. On the eastern slope of the Long Range are stands of commercial forest. Settlements along the coast are somewhat more widely spaced than in other regions. The principal towns include the important fishing centre of Port au Choix and the town of St. Anthony.

The west coast. The west coast region stretches southward from Bonne Bay, at the base of the Great Northern Peninsula, to Cape Ray on the Cabot Strait. The northern section is well forested, while from St. George's Bay to the Codroy Valley is the province's best agricultural land. The principal population centres are Corner Brook, a city based on a large pulp and paper complex, and Stephenville, originally developed in conjunction with a U.S. air base that was later converted into a centre for light industry and also became the site of a pulp and paper mill.

Sherman Hines/Masterfile



François, a fishing village on the rugged south coast of the island of Newfoundland.

The south coast. The most spectacularly rugged section of the island's coastline is probably the south coast, stretching from Cape Ray eastward to Fortune Bay. Here, the land generally rises abruptly from the sea, forming numerous embayments separated by rocky peninsulas. The rivers are short and turbulent, but the head of Bay d'Espoir, the site of a major hydroelectric project, and the rivers and valleys that enter it are well forested. Many small settlements that once dotted the coast have been abandoned, and the population is now centred in fewer but larger communities such as Channel-Port-aux-Basques at the western extreme of the region; it is a major port of entry for mainland Canada.

The Burin Peninsula. Almost all settlement is coastal and directly related to the fishery. The larger towns, Marystown, Burin, Grand Bank, and Fortune, are all major deep-sea-fishing ports and centres for the fresh-fish-processing industry. St. Lawrence produces Canada's total supply of the mineral fluorspar, an important agent in metallurgy.

The island interior. The interior is largely uninhabited. The pulp and paper industry fostered the development of such towns as Grand Falls, Windsor, and Bishop's Falls, while international air traffic made Gander into a major service centre for the whole northeast coast.

Coastal and western Labrador. Labrador is composed of two very different regions. The coast south of the Hamilton Inlet is settled with scattered fishing villages. The area north of the inlet has five permanent settlements with predominantly native Indian and Inuit populations. The largest town is Happy Valley-Goose Bay, which developed around a military airfield. In western Labrador the twin mining towns of Labrador City and Wabush are modern, well-planned, and prosperous; they form the centre of an industry that produces more than one-half of Canada's iron ore. Churchill Falls is the site of one of the world's greatest hydroelectrical developments, supplying power to Quebec and indirectly to the northeastern United States. Game fishing is a major activity.

The people. Considerably less than one-half of 1 percent of the population is aboriginal. A tiny population of Inuit (Eskimo) and North American Innu (Montagnais-Naskapi) occupy four or five villages on the northern coast of Labrador, retaining their original languages and a portion of their ancient cultures. On the island, the indigenous Beothuk are extinct, but a small Micmac community still exists.

Dominance of British peoples

The greater part of Newfoundland's people represent an extraordinarily homogenous group. About 95 percent of them trace their origin to the southwestern counties of England or to the southwestern region of Ireland. A small number trace their ancestry to other English and Irish counties, Scotland, Wales, the Channel Islands, or the coasts of Normandy and Breton in France. More than 98 percent of the people are English-speaking. Historically, social groups have been defined along denominational rather than linguistic or ethnic lines. The three principal denominations are Roman Catholic, Anglican, and United Church of Canada.

From the beginning of settlement, Newfoundland was a staging point for further emigration, and, in the 20th century particularly, there has been a large emigration to the United States and to other parts of Canada. Within the province the trend toward urbanization has been accelerated by government policy to reduce the costs of services by centralizing population. More than half the population is urban, nearly half of which live in the metropolitan areas of St. John's and Corner Brook.

The economy. Newfoundland's economy has traditionally been based on exploitation and export of natural resources. The economy is still resource-oriented although rapidly becoming more complex. Secondary manufacture is growing, and service and communications-related industries have expanded even more rapidly. Historically, the fisheries were Newfoundland's chief industry, and until the early 20th century they were virtually the only industry. The fisheries, although having diminished comparatively as a source of provincial wealth, remain the largest employers of labour and continue to be the main economic base for hundreds of coastal villages and towns. Agriculture is of little significance.

Resources and industry. The mineral resources of the island and of Labrador have become the largest contributor to the gross provincial product. The growth of mineral production has largely arisen from development of the immense iron ore reserves of Labrador, although other minerals, such as copper, lead, zinc, gold, asbestos, and gypsum, are also important. Significant offshore oil and gas deposits are a potential source of wealth. World demand for newsprint and other paper products has encouraged the rapid exploitation of the forests of black spruce, a wood suited to that manufacture.

Energy potential

Newfoundland's power requirements are met chiefly through the capture of its vast hydroelectric resources, particularly those of Labrador. However, the province of Quebec, which lies between these power sources and the markets in the more populous regions of North America, has denied Newfoundland access to transmission routes except under certain stringent conditions. Thus, virtually the entire yield of the vast Churchill Falls development is sold at bargain prices to Quebec and then resold by that province at a profit.

Fisheries and agriculture. Modern technology has changed the nature of the fisheries. For centuries the industry revolved around codfish, caught in inshore waters and salted or dried primarily by small family-run facilities for domestic and international markets. Today, while the inshore fishery, processing chiefly turbot, plaice, and redfish, is still the economic and social backbone of rural Newfoundland, the greatest part of the total catch is processed in large plants located strategically along the coast. Secondary processing is assuming an increasingly important role. Agricultural products, consisting mostly of vegetables, fruits, and livestock, are marketed locally, and most of the foodstuffs consumed must be imported.

Transportation. Until the end of the 19th century, virtually all communication among the coastal settlements was by sea. A narrow-gauge railway was built during the 1880s from St. John's to Channel-Port-aux-Basques along a route that touched the major bays. Subsequently, branch lines were developed to serve population centres and to connect with coastal steamships that carried passengers, freight, and mail to the villages not served by the railway. That railway has been abandoned, and the province now has only a single line that carries freight from Labrador to the Gulf of St. Lawrence. Elsewhere, rail has been replaced by road. A section of the Trans-Canada Highway crosses Newfoundland island more or less following the old rail line, while secondary roads link virtually every settlement on the island. Except for a short section along the Strait of Belle Isle and a gravel highway from Happy Valley-Goose Bay to Esker, most Labrador communities are still without road connections to their neighbours; the construction of landing strips for light aircraft has made possible some regularly scheduled air services. More recently work has been completed on a highway that links Labrador City and Wabush with the road network of Quebec. The island is linked to mainland Canada by ferries operating between Channel-Port-aux-Basques and Argentia on the south coast and North Sydney, N.S. National, international, and regional air service is provided by various lines.

Connections with the mainland

Administration and social conditions. In most ways, administrative organization and social conditions in Newfoundland are typical of those in other Canadian provinces.

Government. Under the terms of union with Canada, Newfoundland is represented in the Canadian Parliament by six senators and seven members of the House of Commons. The provincial government, completely autonomous in those areas reserved to its jurisdiction by the Canadian Constitution, consists of a lieutenant governor representing the crown and a unicameral legislature of 51 members elected by universal suffrage. The government is organized on the British parliamentary model. The executive council, or Cabinet, is chosen from among the elected members of the party representing the majority in the House of Assembly. Between 1949 and 1972 the Liberal Party won every election, with the opposition Progressive Conservative Party unable to win more than a few electoral districts. The Conservatives finally won power in the latter year, supported by a new generation who had not been involved in the great confederation debate and who were disillusioned with Liberal leadership and their policy of resettlement and centralization of many small coastal communities. In the election of 1988 the Liberals were returned to power with a substantial majority of electoral seats, although the Conservatives had won a small majority of the popular vote. The New Democratic Party (NDP) has never been a significant factor in Newfoundland.

Local government was slow to evolve in Newfoundland. Not until 1888 was the first municipal government elected in St. John's, and the next town council was not created until 1938. By the late 20th century, most communities were incorporated as towns or as components of rural district organizations. Elected councils are empowered to levy taxes and, under the aegis of the provincial Department of Municipal Affairs and Housing, are responsible for the management of local concerns.

Education. Newfoundland's educational system is a mixture of state and church activity. Historically, educational services were provided by the churches, which were permitted to retain administrative controls even when the government assumed the major financial burden. The rights thus acquired were enshrined in the provincial constitution, the provincial Department of Education being divided into five denominational divisions. In 1969 the Department of Education was reorganized: school districts were consolidated, school boards were reduced, and denominational services were integrated. The individual denominations, however, continue to maintain their own administrative offices and are responsible for certain functions. Vocational, technical, and trade schools are nondenominational in character and are operated by the provincial government. Founded as Memorial University College in 1925, the Memorial University of Newfoundland, in St. John's, is the province's single university. Nondenominational and managed by an autonomous board of regents, it has become one of Canada's larger universities with professional schools and research programs, some directly related to Newfoundland's physical and sociological environment.

Educational consolidation

Health and welfare. The provincial government administers virtually all health services, which are free to all residents. With few exceptions, members of the medical profession participate in the Medical Care Plan. Newfoundland differs from most other Canadian provinces in that few private agencies exist to provide services to the aged and children and to the variously disadvantaged. Almost all such services are provided by the provincial government. This reflects both the recent development of municipal institutions, the great dispersal of settlement, and the small population of most communities. The police force that serves St. John's is, in theory, responsible for the whole of the province. They are assisted outside St. John's by the Royal Canadian Mounted Police.

The same problems of population distribution and geography also combine to keep the cost of living somewhat higher and wage rates somewhat lower than in other parts of Canada. However, an inherent factor within a standard of living is the somewhat intangible notion of quality of life. Thus a cash income that statistically indicates subsistence below the poverty level might represent a comfortable living to a Newfoundland fisherman owning his own home, fishing equipment, and premises.

Cultural life. The insularity of Newfoundland and the isolation of hundreds of small communities have combined to preserve distinctive cultural characteristics that have long since disappeared in the land of their origin. Thus, a most distinctive feature of Newfoundland is the local speech, which preserves much of the flavour of its West Country origins in 17th-century England. Over several centuries it has been modified only by the influence of the Irish who at one time constituted half the total population. But there is really no such thing as a Newfoundland accent, for there are almost as many dialects as there are bays. The differences between these dialects and "standard English" are not merely in accent but in vocabulary, in sounds, and in syntax.

Folk culture

Though isolation produced distinctive language patterns and usages, it created as well the necessity for local entertainment. This factor, perhaps in combination with the Irish traditions of bards and minstrels, encouraged the development of the art of the storyteller and the preservation and creation of an immense body of folk music and song. Indeed, Newfoundland has been categorized as a living archive of folklore, folktale, and folk song. With a few notable exceptions, Newfoundlanders have not been exceptionally productive in the visual arts. But about 1970 painting and printmaking began to flourish, giving rise to a number of commercial galleries where there had been none. At the same time there was a renaissance of professional theatre, an institution that had nearly succumbed with the advent of television.

A provincial museum and archives in St. John's is maintained under the auspices of the Department of Culture, Recreation, and Youth. Smaller museums and cultural centres are located in other major towns throughout the province. Memorial University also serves as a centre of culture, promoting the arts through a variety of programs.

Recreational activity in Newfoundland is mostly related to the province's abundant and well-preserved natural scenery and wildlife, an endowment that serves both local and tourist interests. The province maintains a generous amount of parkland and preserves, including the magnificent Gros Morne National Park, and its history is reflected in such places as L'Anse Aux Meadows, the only authenticated Viking site in the hemisphere.

Publishing

Publishing was slow developing in Newfoundland. But in recent decades this field has also opened up, based primarily on a resurgence of interest in the local tradition and idiom. St. John's has always been well served by newspapers, although the score or more that appeared during the 19th century and flourished for various periods of time have now been replaced by the daily *Evening Telegram* and two Sunday papers, the *Express* and the *Telegram.* The *Western Star* is published in Corner Brook, and several weeklies serve other smaller communities. Radio broadcasting has been widespread since the 1930s, and the first television station was established on the island in 1955. The two Canadian television networks have stations in St. John's, and cable television is available.

Newfoundland's distinctive culture has begun to recede before the impact of mainstream North American influences, including the increase in the amount of radio and television transmission and the influx of large numbers of professional people. A particular and pressing care is to preserve the distinctive elements of the folk culture as a vital force in the province's life.

HISTORY

The official discoverer of Newfoundland was the Genoese-Venetian navigator Giovanni Caboto, better known to history as John Cabot, who reached the island in 1497 while sailing under the English flag. The Grand Banks doubtless had been known to fishermen earlier, but Cabot's enthusiastic reports opened the way for international rivalries over the region. Early in the 16th century, English, French, Basque, and Portuguese fishermen were contesting for catches; by 1600 England and France were the chief rivals for Newfoundland.

Early restrictions on settlement. Several attempts made at colonization during the 17th century met the hostility of the English fishermen and, after 1634, of the English crown. In 1699 Parliament prohibited settlement of the island except as necessary to maintain the cod fishery. This remained the basic governing directive for Newfoundland for more than 150 years, during which time the island became virtually a private fiefdom ruled by commercial fishing interests.

The rationale for this policy was twofold. First, under the prevailing concepts of international trade, the fishing of the Grand Banks remained a mother-country industry rather than that of a colony, with definite economic advantages accruing to England. Second, the region was looked upon as a "nursery for seamen," a training ground for the crews that were to maintain England's maritime superiority into the distant future. This latter idea also led the French in 1662 to found Placentia on the southeast coast as a base of operations. From Placentia the French destroyed every English settlement on the island in actions that took place in 1696, 1705, and 1708. Treaties in 1713, 1763, and 1783, the latter marking the armistice ending the American Revolution, recognized British sovereignty but granted French fishermen the right to land and dry their catches along portions of the northern and western coasts. This "French shore problem" remained a thorny one for the islanders until 1904.

French raids

During the 17th and 18th centuries the island's population increased despite repressive legislation, the denial of property rights and social services, and a particularly vicious judicial system managed by the captains of fishing vessels. Britain found it necessary to appoint a naval governor in 1729 and to establish a proper judicial system in 1792. Both governor and court remained only for the summer fishing season, however, and not until 1824, with the appointment of a resident governor and council, was a settled colony finally acknowledged.

In 1832 local agitation led to the addition of a popularly elected assembly. In 1855 the island was granted full responsible government after a campaign that reignited a bitter underlying sectarianism. As the hold of the transient fishermen over the island's life weakened toward the end of the 18th century, many of the prosperous fishermen settled in as merchants, and government office-holders increased in number. Large groups of labourers were imported for the cod fishery, 11,000 coming in the peak year of 1814–15 alone. These immigrants, mainly Irish Catholics, continued to stand for the rights of fishermen and servants against the new and largely Protestant middle class, creating a sectarian spirit that persisted in Newfoundland's political life well into the 20th century.

Toward confederation. Between 1864 and 1869, Newfoundland was involved in discussions leading to Canadian confederation. The idea of confederation, however, was rejected at the polls both then and after negotiations were reopened in 1895. New radical social policies were heralded with the formation of the People's Party by Edward Morris in 1907 and the Fishermen's Protective Union by William Coaker in 1908. The political rivalry of Morris and Coaker was submerged in the common efforts of World War I.

For nearly 150 years, rights to the coasts of Labrador, an ill-defined geographic entity, were a hotly debated issue. By 1826, British tribunals had given Newfoundland jurisdiction over the Atlantic face, Canada over the coasts on the Gulf of St. Lawrence. However, the inland boundaries and administration of Labrador were to remain unsettled until 1927, when Newfoundland was awarded the huge inland region in a definition that said that the "coasts" reached to the headwaters of rivers emptying eastward into the Atlantic Ocean.

The depression that afflicted the Western world after 1929 completely overwhelmed Newfoundland's already struggling economy, which had been overtaxed by the war and by the cost of building a transisland railway. In 1934 Newfoundland's government was reformed as a commission appointed by and responsible to the government of the United Kingdom. The commission was finally able to reestablish the colony's finances on a firmer foundation.

Confederation contested

Following World War II a national convention was elected to recommend possible forms of future government to be voted upon by the people. The convention failed, however, to recommend anything beyond a continuation of the existing government or a return to the pre-1934 constitution. The British government, nevertheless, added the confederation question to the ballot. The first of two referenda held in 1948 was inconclusive; the second gave a clear majority to confederation. On March 31, 1949, Newfoundland entered the Canadian Union.

Through the 1950s and '60s Newfoundland underwent changes as modern technology transformed industry and the nature of employment. Family-run operations were replaced by large fisheries and canneries, and a steady migration took place from smaller villages to the industrial towns. Population growth was rapid. Economic recession in the 1970s and '80s affected the province, although the impact of high unemployment was somewhat offset by social programs. The discovery of rich offshore oil and gas fields promised new wealth. (L.H./Ed.)

Nova Scotia

One of the four British colonies federated into the Dominion of Canada in 1867, Nova Scotia comprises the peninsula of Nova Scotia, Cape Breton Island (separated from the mainland to the southwest by the narrow Strait of Canso), and a number of small adjacent islands. The total area is 21,425 square miles (55,490 square kilometres). Along the 17-mile- (27-kilometre-) wide Chignecto Isthmus, which seems to thrust the peninsula into the Atlantic Ocean, runs the province's only land boundary, with New Brunswick to the west. Two arms of the Gulf of St. Lawrence, Northumberland and Cabot straits, separate it, respectively, from Prince Edward Island on the north and Newfoundland on the northeast. To the east lies the Atlantic and to the southwest the Bay of Fundy. Halifax is the capital.

Nova Scotia—with New Brunswick and Prince Edward Island—is one of Canada's Maritime Provinces, and both its past and its present are tied closely to the maritime life of fishing, shipbuilding, and transatlantic shipping. It was the site, in 1605, of the first permanent North American settlement north of Florida, established by the French. Among the legends that pervade the province is that told by the U.S. poet Henry Wadsworth Longfellow in his tale *Evangeline,* an account of the deportation in 1755 of the French inhabitants by a fearful British governor. The province's contemporary life retains some of the feeling of 17th-century Acadie, or Acadia, the Micmac Indian name adopted by the French for the region before Scottish colonists implanted their own name of Nova Scotia.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief, soils, and drainage.* Nova Scotia's five upland regions reach a maximum height of more than 1,700 feet (520 metres) above sea level in the Cape Breton Highlands. The most important lowlands lie along the Bay of Fundy and the Minas Basin, in the southwest and along the Northumberland Strait. Many of the tens of thousands of acres of marshland in the west created by the tremendously high tides of the Bay of Fundy have been turned to agricultural use by dikes, which were begun in the early 18th century by the Acadian French.

Nearly 90 percent of Nova Scotia's landmass is unsuitable for agriculture. Most of the southern peninsula rests upon acidic granite, and a large part of Cape Breton Island is mountainous, forested terrain of acidic and metamorphic rock. Podzolic soil dominates, with some gray wooded soil mainly along the Northumberland Strait. In a few pockets where clastic sedimentary rock, mostly sandstone, underlies the soil—as in the Annapolis Valley, along parts of Northumberland Strait, and at Cobequid Bay—the land supports orchards and field crops.

More than 3,000 lakes and hundreds of short rivers and streams either have been impounded by, or have cut

K. Wright/Miller Comstock Inc.

Ships docked in Halifax Harbour at Dartmouth, N.S.

through, the irregularly high and low landscapes. The best known of the lakes, Bras d'Or on Cape Breton Island, which is saline, is connected to the Atlantic through three short channels. Many intruding heads of land make the lake's 424 square miles a geographic complexity.

Climate. Over its 350-mile length, Nova Scotia has a modified continental climate, comparable to that of northern Europe. Rarely does the temperature exceed 88° F (31° C) or fall below 14° F (−10° C) anywhere in the province. At Halifax International Airport the average July temperature is 65° F (18° C), and the average January temperature is 23° F (−5° C). Yearly precipitation at that station averages about 59 inches (1,500 millimetres), including some 107 inches (272 centimetres) of snow. The southwestern and southern shores of Nova Scotia have both milder and wetter climates than the rest of the province. Rainfall varies from 55 inches (1,400 millimetres) in the south, where fog may occur on as many as 90 days, to 40 inches elsewhere.

Plant and animal life. Forests occupy about 78 percent of the land area. About one-fifth of the woodland is held as crown land. Softwoods are by far the most numerous, led by species of balsam, spruce, hemlock, and pine; birch and maple make up most of the hardwoods. Animal life includes such game species as deer and moose and, among the birds, partridge, pheasant, and duck. Trout and salmon are common freshwater game fish; tuna is a favourite of deep-sea fishermen.

Settlement patterns. Subsistence living on family farms has been characteristic of Nova Scotia agriculture. Along the coasts many families living on marginal land have combined farming with fishing for lobsters and groundfish; in other areas farming and lumbering are often combined. Mixed farming and dairying are carried out in the fertile lands of Guysborough and other eastern counties, as well as on the Chignecto Isthmus and along the Fundy shore. Though apples, grapes, and pears are also found in other sections, the Annapolis Valley orchards, dating back three centuries, are the most productive. More efficient land use and increased mechanization have reduced the number and increased the size of farms while improving production and income. About 6 percent of the people live on farms, and another 40 percent are nonfarm rural dwellers.

In the 19th century, as steamships replaced schooners, the outports dwindled. Coal and steel and textile industries drew workers to Sydney and Halifax, but some smaller towns continued enterprises like boatbuilding, woodworking, and food processing. Internal population movement has been toward the Halifax–Dartmouth area, where there is employment in the shipyards, naval dockyard, construction industries, and assembly plants.

The people. When Europeans arrived in the 16th and 17th centuries, Micmac Indians greeted them. This seminomadic tribe ranged over the Maritime Provinces and into the Gaspé Peninsula and later spread to Newfoundland and New England. Their Algonquian language is reflected in such Nova Scotia place-names as Musquodoboit, Pugwash, and Shubenacadie. Of 14,000 Micmac Indians in Nova Scotia, more than 3,000 live on reservations.

Ethnic components of the population

About one-eighth of Nova Scotia's people are descended from the Acadian French, who were allowed to return from exile after the British took French Canada in 1763. Most of the remaining seven-eighths are descended from settlers from the British Isles, both English and Scottish. Acadian communities are now located around Yarmouth along St. Mary's Bay and on Cape Breton Island. New England planters resettled the Minas Basin after 1755, and British Empire loyalists founded Shelburne in 1783 and helped to populate Halifax. Yorkshiremen settled in Halifax county. Scottish Highlanders tended to settle in the Northumberland Strait counties and Cape Breton Island. Also in the 18th century some Ulstermen, including migrants from New Hampshire, established farms in the Truro and Onslow districts. Later Irish came in family groups by way of Newfoundland and directly from Ireland and settled in and near Halifax. German immigrants in the 1750s founded the seaport of Lunenburg, and others came a century later. Other minorities include small numbers of Dutch, Italians, and Hungarians, who have immigrated since World War II. More significant are the black communities near Halifax and Shelburne, dating from the days of the West Indian slave trade.

English is the only spoken language of more than 90 percent of the people. Few Nova Scotians remain who have French as their only spoken language, but nearly 8 percent of the people are bilingual. More than one-third of the people are Roman Catholics; among the Protestant denominations the United Church of Canada is the largest, followed by Anglicans and Baptists. Since 1960 immigration into Nova Scotia has not been numerically significant, but emigration of both the undertrained and the well-educated has continued.

The economy. Nova Scotia's diversified economy still relies heavily on maritime activities. It embraces some international and many local resource-based manufacturers housed in industrial parks. Essential to industrial development have been such crown corporations as the provincial Industrial Estates Limited and the federal Cape Breton Development Corporation. The federal Atlantic Canada Opportunities Agency assists with incentive grants mainly for small and medium-sized businesses. In addition, federal equalization payments help to compensate the province for above-average unemployment rates and substandard total and per capita incomes. Most of the labour force is employed in service industries.

Farm products

Agriculture and fishing. Farming has focused on dairy products, livestock, poultry and eggs, and fruit. Extensive forestry resources supply large pulp and paper mills, numerous sawmills, and expanding industries in Christmas trees and maple syrup.

The catching, processing, and exporting of fish continue as major Nova Scotian industries. Canada's stock-conservation and surveillance policies in the internationally recognized 200-mile coastal zone ensure that Nova Scotia's fisheries will remain productive. For fishermen from 250 harbours, lobster and cod are the most lucrative catches, but scallops, haddock, and herring are also important. Aquaculture has shown considerable promise.

Industries. Mining is another major industry in Nova Scotia. Coal is the leading mined product, and some 70 percent of the province's energy needs are met through thermal coal. Salt and anhydrite production also meet a wide demand, and the provincial gypsum deposits yield 75 percent of Canada's supply. There are also significant supplies of barite and construction materials, such as sand and gravel. A profitable tin mine has operated in East Kemptville since 1986. Upgraded mills in Sydney and Pictou manufacture selected steel products. Near Sable Island tested reserves offer promise of large quantities of oil and natural gas for export. The continent's first tidal energy project, completed in 1984 near Annapolis Royal, harnesses the tides of the Bay of Fundy to enhance the province's hydroelectric energy output.

More than one million tourists annually stimulate Nova Scotia's economy with cash and jobs generated for about 7 percent of the work force. Also significant for the provincial economy are four Canadian Forces bases: Halifax, Cornwallis, Shearwater, and Greenwood.

Transportation. Shipping remains a major enterprise in Nova Scotia. Point Tupper accommodates the world's largest oil carriers, and Halifax, a railroad terminus and year-round port, has facilities for all types of vessels, including huge container ships. Other transportation needs are served by a network of paved highways, by a rapidly growing trucking industry, which is partly displacing rail service, and by an international airport at Halifax and several smaller airports. Ferries operate between Nova Scotia and ports in New Brunswick, Prince Edward Island, and the U.S. state of Maine.

Branches of government

Administration and social conditions. *Government.* Nova Scotia's governmental structure is similar to that of all Canadian provinces. A lieutenant governor appointed by the federal government serves as representative of the crown. The premier, who is the leader of the party in power in the provincial parliament, selects his cabinet from among his colleagues. The judiciary is federally appointed, while government employees are recruited through a nonpolitical civil service. Provincial income is derived from two virtually equal sources: the federal government and various provincial taxes and fees.

Since confederation there have been two major political parties in Nova Scotia, the dominant Liberals and the Conservatives (now the Progressive Conservatives); the New Democratic Party has had little success.

Education. The nondenominational system of compulsory, free public education, dating from the mid-19th century, consolidated the many rural school districts into fewer, larger, and more efficient districts during the 1970s. Considerable emphasis is placed on commercial and vocational training at the secondary as well as the college level. Dalhousie University (founded 1818) in Halifax is perhaps the best known of the institutions of higher learning, though a number of Roman Catholic and other private institutions offer diversified curricula. Agriculture, social work, and teacher training are among the specialized areas served by specific institutions. St. Francis Xavier University, in Antigonish, has attracted international interest in its adult-education programs.

Health and welfare. In 1969 Nova Scotia joined the federal Medical Care Program. The premium is paid from the province's general revenues, and insured services include all medically necessary procedures. The province provides facilities for mental health, dental care, tuberculosis control, and other public-health services. Nova Scotia's welfare services are similar to those in other provinces, covering old age assistance, allowances for the blind and disabled, social assistance, social development, child welfare and adoptions, and service to unmarried mothers. The Department of Public Welfare operates institutions for mentally retarded children and for delinquent youth and supervises some private institutions.

Cultural life. Its geographic position has kept Nova Scotia somewhat removed from the mainstream of Canadian life, centred in Ontario and Quebec. The improvement of transportation, however, and the growing impact of the nationwide communications media have brought the province closer to the mainstream, enhancing a modern life-style in the province and eroding some features of traditional Nova Scotian life. Nevertheless many areas of the mainland peninsula and of Cape Breton Island retain an unretouched image of the past.

Preservation of traditional cultures

Scottish culture is particularly vigorous in Nova Scotia. Pictou county alone has several bagpipe bands, while St. Francis Xavier University offers courses in Celtic studies, and the Gaelic College in St. Ann's, Cape Breton, fosters piping, singing, dancing, and handicrafts. Clan gatherings take place annually at the Gaelic College to celebrate the Gaelic Mod, a festival of Highland folk arts. Each summer individual and community arts are taught at Tatamagouche, and an Acadian festival is held at Clare. Other popular attractions include the annual Nova Scotia Military Tattoo and the Atlantic Winter Fair.

The Neptune Theatre and the Art Gallery of Nova Scotia are major cultural assets. Halifax is the centre of repertory drama, and the universities have their own theatre productions. Resident and visiting painters work in such south shore communities as Mahone Bay, Lunenburg, and Peggy's Cove. Woodcarving, pottery making, hooking, and weaving are also pursued.

Impressive historic sites include the restored fortress of Louisbourg, Champlain's habitation at Port Royal, and the Halifax citadel. Also of interest are the Breton Highlands National Park and the Alexander Graham Bell National Historic Park and a museum at Baddeck and the Grand Pré memorial in the Annapolis Valley. Provincial parks are under continual development, and accommodations and recreational facilities are increasing.

HISTORY

In the 17th and 18th centuries Nova Scotia experienced instabilities of colonization, struggles for power originating in the rivalries between London and Paris, and migratory and military pressure from the colonies to the south that were to become the United States. The territory passed back and forth between France and England until 1713, when the French, retaining Cape Breton Island and other areas, began construction of the powerful Louisbourg fortress. Halifax was founded in 1749 as a counterbalance and populated with some 4,000 British settlers. In the 1750s the French Acadians, who refused to swear allegiance to the British crown, were expelled.

Offers of free land attracted immigrants from the British Isles and New England; the latter group helped to secure the first elected assembly. By the time of the American Revolution, New Englanders constituted roughly one-half of Nova Scotia's population; though they tried to remain neutral, four delegates attended the Continental Congress in Philadelphia. During the Revolution and after, some 35,000 loyalists immigrated to the province. Meanwhile, Prince Edward Island had split off from Nova Scotia in 1769, and New Brunswick followed in 1784; Cape Breton Island was reunited with Nova Scotia in 1820. In 1848 Nova Scotia became the first British colony to exercise the prerogative of government responsible to the people through their elected representatives. Despite economic and political opposition, confederation with Ontario, Quebec, and New Brunswick was carried out in 1867.

The Prince Edward Island and Nova Scotia split

As a separate British colony, Nova Scotia prospered from its forestry, fisheries, and shipbuilding for the first two-thirds of the 19th century. Under the Reciprocity Treaty of 1854, the north–south flow of commerce and Nova Scotia's normal market and supply source in New England seemed secure. Nova Scotia benefited further from the increased demand for duty-free natural products during the U.S. Civil War. The Canadian tariffs on manufactured goods, the alleged pro-Confederate sentiments of Britain and Canada, and the protectionist pressures of some states in the victorious North led to the nonrenewal of reciprocity in 1866 and the levying of further protective tariffs by both nations. Canada also lost the preferential treatment by Britain previously granted under colonial policy. Railways, meanwhile, were changing continental traffic patterns, and Canada and the U.S became rivals for the east–west inland trade. Canada's tariffs, intended both to protect its own manufactures and to assist in financing its railways from coastal ports to Central Canada, did not assure the province of year-round commerce, because transportation costs to the distant population centres were so high. Nova Scotia's prospects were further set back as iron steamers replaced the wooden sailing vessels that had been the pride and chief industry of the province. The steamers usually bypassed Nova Scotian ports for those in the United States, and even the lucrative trade with the West Indies dwindled. Also, the opening of the western provinces to settlement tended to drain the older regions of some vigorous elements of the population and to deprive them of investment capital and other resources.

By the last decade of the 20th century Nova Scotia's diversified economy and extensive natural resources had the potential to expand business and reduce unemployment. The 1989 Canada–United States free-trade agreement was regarded as a benefit to the province's fisheries and to its shipping and energy sectors, and continuing federal assistance was a vital factor in the maintenance of social and economic programs. There was, at the same time, a general agreement among the political and social institutions that Nova Scotia should continue to advance technologically but with a minimal sacrifice of traditional human values.

Agreement on technological growth

Prince Edward Island

The original Micmac Indian metaphor Abegweit—popularly translated "Cradled on the Waves"—aptly describes the slender crescent of land in the Gulf of St. Lawrence, known since 1799 as Prince Edward Island (P.E.I.). During the French regime (1720–58) it was called Île Saint-Jean, but when the British took over they first Anglicized the name to Saint John's Island, then attempted to call it New Ireland, and finally named it for the Duke of Kent, commander of the British forces in North America. In 1867 Prince Edward Island became the seventh, and smallest, province of Canada. The name of its capital, Charlottetown, commemorates the wife of King George III.

Curving from North Cape to East Point, the island is about 140 miles (225 kilometres) long, ranging from 2 to

40 miles (3 to 65 kilometres) in width. It lies between 46° and 47° N latitude, and 62° and 64° W longitude. On the south the Northumberland Strait separates the island by at least nine miles from the mainland provinces of Nova Scotia and New Brunswick. There are three counties: Prince, Queens, and Kings. The land area is 2,185 square miles (5,660 square kilometres).

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief and soils.* Numerous streams, bays, and tidal estuaries indent the irregular coastline. On the north side the bays are generally blocked by dunes, but on the east and south there are good natural harbours. Along the 1,100 miles of shoreline, red sandstone cliffs average about 20 feet (6 metres) in height but occasionally exceed 50 feet (15 metres). Most of the island is within 5 miles of the sea or a tidal inlet, and no spot is more than 10 miles from salt water. Water pollution has not been extensive, but effluents from processing plants, pesticides, erosion, oil leakage, and sedimentation have seriously degraded some water resources.

Bill Brooks/Masterfile

Red sandstone cliffs along the Gulf of St. Lawrence at Cavendish Beach, on the north coast of Prince Edward Island.

The landscape ranges from rolling hills in central Queens county to level stretches in western Prince county. The highest elevation is 466 feet (142 metres) above sea level in Queens county. The soils, classed as podzols, rest on a sandstone base and are low in plant nutrients and high in acidity. On more than half the island, the red, sandy loam is suitable for cultivation.

Highest elevation

Climate and plant and animal life. The province has a relatively mild climate. The average mean temperature is 42° F (6° C), and the mean annual precipitation is 40 inches (1,016 millimetres).

Forests cover two-fifths of the island. The original stands were mainly deciduous, but coniferous trees are now much more common. Besides the native birch, maple, pine, and spruce, foreign species include linden, horse chestnut, black walnut, and European mountain ash. Tall trees line town streets as well as country lanes.

Prior to European settlement, moose and caribou roamed the island, as did black bears up to the early 20th century. Wildcats were exterminated, but foxes and snowshoe hares endured, and the nearly depleted beaver has been reintroduced. Mink, weasels, and muskrats are numerous. The coyote and the prolific skunk are unwelcome invaders. The ruffed grouse is a native bird, but ring-necked pheasants and Hungarian partridges are imports. Canada geese and brant are regular migrants. Cormorants are increasing, while black ducks are diminishing. Island rivers support such fish as trout, perch, and salmon. There is an active wildlife conservation program.

Settlement patterns. Despite its largely rural character, Prince Edward Island is the most densely populated province in Canada. Population growth has been most apparent in the urban areas of Charlottetown and Summerside. The smaller towns, such as Souris, Kensington, and Tignish, have exhibited little change. Family farms are increasing in size but decreasing in numbers. Kings county is the least populated area.

The people. More than three-fourths of the residents are descendants of early settlers from the British Isles: Highland Scots, Englishmen, and both southern and Ulster Irish. Other ancestral strains include several hundred loyalists, who settled after the American Revolution, and about 30 Acadian families, recorded in the census of 1765, who are the progenitors of several thousand present-day French-speaking island residents. Many original settlement patterns can still be traced in place-names and family names. The Acadians mostly live south of Rustico, near Cape Egmont, and west and north of Cascumpeque Bay. Since 1945 a number of Dutch families, as well as a sprinkling of Americans, Belgians, Lebanese, and Germans, have settled in Prince Edward Island.

The Acadians

Roman Catholics make up the largest religious denomination, followed by United Church members, Presbyterians, Anglicans, and Baptists. English is the predominant language of the island; relatively few residents speak only French. Bilingualism is encouraged, however, by an optional French immersion program in the schools, and classroom instruction in French is available.

The economy. A fertile soil and temperate growing season favour agriculture; the island's location makes it a base for lobster, shellfish, and groundfish (cod, hake, flounder, and redfish) operations; and its sandy beaches, warm waters, and other tourist amenities attract hundreds of thousands of summer visitors. Despite the success of these major industries, the provincial economy has lagged appreciably behind national averages for productivity, employment, and per capita income. Limited resources, distance from large markets, high transportation costs, and the nation's highest electric-power rates all hamper industrial enterprise.

During the late 20th century, federal-provincial agreements enabled the province to institute a series of reforms aimed at assisting the people to create viable economic enterprises for themselves. These included full exploitation of agriculture; development of tourist facilities; better use of forest assets; improvement of fisheries; expansion of manufacturing; public investment in housing, health, and welfare services; and the extension of programs of education and training. The reforms resulted in upgraded living standards, but with the growing dependence on government assistance and a strain on provincial resources.

Agriculture and fishing. Agriculture continues to serve as the economic base. The traditional primary crop is potatoes; turnips, hay, and grains are also grown. Dairying is a major industry, and there is some cattle and hog raising. Tobacco is grown in the eastern part of the island, and the raising of fur-bearing animals is pursued in the western part. Private woodlots yield both hardwood and softwood for lumber, pulpwood, and fuel.

Fishermen depend primarily on the lobster catch, but oysters, clams, and scallops are also important. A secondary industry is the harvesting of Irish moss (a seaweed) for its carrageenan, an extract with several commercial uses. Also lucrative are catches of cod, haddock, tuna, and mackerel. Onshore plants process and freeze sea products; several aquaculture enterprises are established.

Irish moss production

Industry. Industrial development is encouraged by the P.E.I. Lending Authority. Shipping frozen fried potatoes, green vegetables, and berries is now a leading business activity. Food processing has evolved from many small local plants into a few large enterprises. More than 600,000 visitors each summer contribute to the island's economy and social life.

Transportation. Transportation problems have always challenged islanders. Cars and trucks having effectively displaced trains as passenger and freight carriers, the Canadian National Railway in 1989 abandoned its rail services on the island. A year earlier a plebiscite favoured the construction of a fixed crossing linking the island and New Brunswick. Regular ferry services between the island and the mainland provinces are maintained, and Canadian Airlines and Air Canada fly daily routes between Charlottetown, Toronto, Montreal, and centres in the Atlantic region. Potatoes, pulpwood, and other products are shipped from ports on the island's south and east coasts.

Administration and social conditions. The provincial government consists of the lieutenant governor and an Executive Council chosen from the 32 members of the Legislative Assembly. The premier heads the ministerial Executive Council, whose responsibilities include education, agriculture, fisheries, tourism, industry, natural resources, and health and welfare.

The province is represented in the federal Parliament by one member from each of four electoral districts: Cardigan, Hillsborough, Malpeque, and Egmont. Four Senate appointments also are allotted to the province.

The Department of Education administers public education from primary grades through senior high school by means of five regional school boards.

Holland College, an institute of applied arts and technology, and the University of Prince Edward Island were both chartered in 1969, replacing two institutions that originated more than a century earlier, Prince of Wales College and St. Dunstan's University. The main campuses are in Charlottetown, but extension courses are offered in other island communities. In 1983 the Atlantic Veterinary College was established within the provincial university.

The Health Services Commission administers the provincial plan for universal medical care. Provincial agencies govern housing and pollution control.

The Charlottetown cultural centre

Cultural life. Charlottetown is the main centre of culture. Both amateur and professional productions are presented at the Confederation Centre of the Arts, which adjoins historic Province House in that city. This complex contains a large auditorium, art gallery and museum, children's theatre, various studios, shops, and a public library.

Other communities benefit from the Prince Edward Island Museum and Heritage Foundation, with its network of regional facilities and cultural programs. The public library system has community branches and a mobile service. Several art galleries and small theatres operate outside Charlottetown. A multicultural festival is presented annually, and there are regional folk festivals, exhibitions, plowing matches, fiddling contests, and highland games. The annual music festival attracts thousands of participants from the schools of the province. The Evangeline district has a reconstructed Acadian Village, a cultural centre, and a museum. Throughout the island such handicrafts as weaving, pottery making, and woodworking are encouraged.

Facilities and programs for sports and recreation are abundant. Every sizable community has an ice-skating rink; and golf courses, baseball and soccer fields, and basketball courts are scattered throughout the island. All major sports are organized into leagues, and island athletes compete in regional and national meets. Harness races are run year-round and culminate in the Gold Cup and Saucer event in August.

HISTORY

Before the European explorations, Micmac Indians from the mainland used the island for fishing, hunting, and some planting in the warmer seasons. Present-day descendants of these Indians subsist mainly on a small island in Malpeque Bay on the north shore.

Legend suggests that John Cabot, the English-sponsored Genoese-Venetian explorer, may have seen the island in 1497, although historians credit its discovery to Jacques Cartier, the French navigator, in June 1534. Claimed for France in 1603 by Samuel de Champlain, the first governor of French Canada, it was not colonized until 1720, when 300 settlers from France established Port la Joie at the entrance to the harbour of Charlottetown. In addition, fishermen and trappers from the French-speaking mainland colony of Acadia established several other small communities on the island. The French regime lasted only 38 years until the British occupied the island in 1758, dispersing more than 3,500 of the settlers.

Awarding of lots to proprietors

The island was formally ceded to Great Britain in 1763, and six years later the colony was separated from Cape Breton and its first governor was appointed. Under British administration the island was surveyed and divided into three counties, each with a township and "royalty," and 67 lots or townships of about 20,000 acres each. In 1767, proprietors, who were expected to promote settlement, were awarded 64 of these lots by ballot, and for the next century absentee-landlord problems beset the colony. Representative government was granted in 1851, and the Land Purchase Act of 1875 ended the controversial land tenure system. By 1900 the population reached 100,000.

In 1864 a conference called to discuss the Maritime Provinces' union prepared the way for the confederation of all the Canadian provinces. This Charlottetown Conference was the forerunner of the Quebec Conference of 1867, which actually resulted in the founding of the Dominion of Canada. Prince Edward Island has thus been known as the "Cradle of Confederation," even though it did not finally join the union until 1873, when forced to do so by severe financial troubles. Resourceful politicians then persuaded the federal government to make concessions that enabled the province to purchase the lands still held by foreign proprietors and resell them to resident farmers; to assume the debt of the island railway; and to obtain assurance of continuous communication with the mainland. Despite these advantages of confederation, the economy of Prince Edward Island remained chronically depressed and dependent on federal assistance through most of the 20th century. (B.O.'G.)

CENTRAL PROVINCES

Ontario

Ontario, the second largest province of Canada in area, occupies the strip of the Canadian mainland lying between Hudson and James bays, on the north, and the St. Lawrence River–Great Lakes chain to the south. Bordered to the east by the province of Quebec and to the west by Manitoba, it covers a total area of 412,581 square miles (1,068,580 square kilometres) and is the most populous Canadian province, representing more than one-third of Canada's total population.

Ontario is also the nation's wealthiest province, having a substantial share of the country's natural resources and its most mature and diversified industrial economy. It is at once Canada's economic pacemaker and a major force in national politics. Some Ontarians call it the "Empire Province," while to those Canadians living outside its boundaries, its preeminent position and the influence of Toronto, the provincial capital, and Ottawa, the national capital, have constituted a not infrequent source of regional resentment.

PHYSICAL AND HUMAN GEOGRAPHY

The forests of the north

The land. *Northern Ontario.* Ontario is composed of two regions of widely different character: northern and southern Ontario. Northern Ontario, as usually defined, lies north of a line drawn from the confluence of the Mattawa and Ottawa rivers southwest to the mouth of the French River on Georgian Bay. Most of the region's 350,000 square miles is part of the ancient Canadian Shield, characteristically marked with a profusion of lakes and rivers, muskeg (bogs), and densely forested, rocky, and rugged terrain. A low plateau, it is generally no more than 1,500 feet (450 metres) above sea level, although it contains the highest point in the province, Ishpatina Ridge, which rises to 2,274 feet (693 metres) west of the Montreal

River. Although the region has always been an obstacle to transportation and communication, its rich mineral deposits, huge forest reserves, and the hydroelectric power potential of its swift rivers have made it a major source of the province's contemporary wealth.

The vegetation is that of the boreal (northern, mountain) forest and includes the black and white spruce, jack pine, tamarack, poplar, white birch, and balsam. The soils of the region, apart from peat, consist largely of brown podzolic (mineral-covered, leached) soils unsuitable for agriculture, except for two clay belts in the Timiskaming and Cochrane farming areas. The region contains parts of two major drainage basins—that of the Great Lakes to the south and of Hudson Bay to the north—separated by a band of higher land running from Lake of the Woods to Kirkland Lake. Major rivers of the northern system are the Severn, the Winisk, and the Albany, while the major rivers in the southern system are the Ottawa and French. Both Hudson Bay and James Bay are bordered by sparsely forested lowlands less than 500 feet above sea level, and, at the northern limit of the province, there is a band of tundra along Hudson Bay.

The climate varies from that of the districts close to the Great Lakes, which are frost-free on more than 100 days a year, to the harsh climate of the Hudson Bay area, where the frost-free period may be as short as 40 days. At Thunder Bay on Lake Superior the mean temperature in January is 5° F (−15° C) and in July 64° F (18° C), the annual precipitation is 28 inches (712 millimetres), and the annual snowfall 84 inches (213 centimetres).

Southern Ontario. Covering less than 10 percent of the area of the province, southern Ontario contains land of gentle relief; its lowest area on the Ottawa River is only 150 feet above sea level, and its highest point—on Blue Mountains south of Georgian Bay—is just over 1,770 feet. The east is divided from the rest of the region by an extension of the Canadian Shield, known as the Frontenac Axis. This crosses the St. Lawrence River east of Kingston and forms the Thousand Islands region. Along the southern edge of the shield lies a series of beautiful lake districts—including the Muskoka Lakes, the lakes of the Haliburton Highlands, and the Rideau chain—which are the province's best-known resort areas. The most dramatic feature of the landscape is the Niagara Escarpment, running from Niagara Falls to the Bruce Peninsula; roads and rail lines pass through its notched valleys, and a nature trail runs along its length. The landforms of southern Ontario were shaped by glacial action, and most of the region consists of gently rolling plains. Both the Ottawa–St. Lawrence lowlands of eastern Ontario and the lands at the western tip of the Ontario Peninsula are, however, quite flat. Glaciers left over most of the region a fertile, gray-brown podzolic soil, although sand plains are also found north of Lake Erie and along the eastern Lake Ontario shore. Rivers of the region are short, draining into the Great Lakes from the Western Ontario Upland and from the Oak Ridge Moraine north of Lake Ontario. Eastern Ontario, however, is drained chiefly by tributaries of the Ottawa River.

The Niagara Escarpment

The original natural vegetation of the area consisted of hardwood forests with great stands of white and red pines on the lighter soils, but during the 19th century land clearing and lumbering removed most of the original forest cover. Since the 1880s, the provincial government has engaged in farm woodlot and reforestation programs.

The climate of the whole region is generally favourable to agriculture, although considerable local variation exists. The eastern section, away from the moderating influence of the lakes, tends to be cooler and more humid than the southern and southwestern zones. Ottawa receives 34 inches of rain and 86 inches of snow yearly, as compared to Toronto's 31 inches of rain and 54 inches of snow. The mean temperature in January for most of the southern region is about 25° F (−4° C), and for July, about 72° F (22° C). The Ottawa valley, however, has means of 13° F (−11° C) and 69° F (21° C).

Pollution. In both regions of the province, industrialization and urbanization have created problems of pollution, the most acute of which are the polluted waters of Lake Erie and the polluted air of the Toronto urban complex. Air and water pollution associated with the mining and pulp and paper industries of the north has also emerged, and increasing concern has been expressed about the presence of mercury in some northern lakes and rivers. The resort region is also endangered because of the high concentration of cottagers.

Settlement patterns. In northern Ontario, settlement has little agricultural base and is largely connected with major industries and transportation routes. Thunder Bay is located at the head of the Great Lakes navigation system and is the transshipment point for western wheat. Sudbury is the centre of a major mining area, as are such communities as Timmins, Kirkland Lake, and Geraldton. Sault Ste. Marie is both an important lake-navigation port and a centre of large steel and paper industries.

Urban growth

Agricultural settlement is more frequent in southern Ontario, where many farms are family-owned. Fields and townships are laid out in a square grid pattern, with roads one mile apart. In areas of French settlement, however, the long, narrow fields typical of French Canadian strip farming may be seen. Villages originally grew up at water-power sites, convenient distribution points, and around early garrison centres. Kingston, the first important town, combined these advantages.

Urban growth has been confined almost entirely to the south-central and southwestern parts of the province, except in Ottawa, the national capital, in the east. The Toronto metropolitan complex—the so-called "Golden Horseshoe"—sprawls along the Lake Ontario shore from Oshawa to St. Catharines and includes the major port and industrial city of Hamilton. Metropolitan Toronto is Canada's largest city, with a hinterland that embraces not only much of the province but a good part of the country. It has a phenomenal rate of growth. Other important urban concentrations in western Ontario include Windsor, London, and Kitchener-Waterloo.

The people. Until the end of the War of 1812, Ontario was peopled chiefly from the United States, by loyalists, frontier farmers, and Quakers and Mennonites from Pennsylvania—the latter forming a nucleus for German settlement in Waterloo County. For the remainder of the 19th century most of the immigrants were Protestants from the British Isles, although both Irish and Scottish Catholics came in large numbers. The first wave of British immigration, between 1815 and 1850, altered the original American character of the province. The second phase of European immigration, from 1896 to 1914, included sizable numbers from Germany, Scandinavia, Russia, Poland, the Ukraine, and Italy. Their arrival coincided with the first great mining discoveries in northern Ontario, and, as a result, the composition of that region's population became much less British in character than the remainder of the province. After World War II, Ontario's immigrants were mainly attracted to the industrial areas of the province, being drawn by the demands of a rapidly expanding economy. Although people from the British Isles still formed the largest single group, they were no longer in the majority, and the cities of Ontario became more cosmopolitan. Roman Catholics have become the largest single religious denomination; they are followed by members of the United Church and Anglicans.

Immigration from Europe

There are no Inuit (Eskimo) indigenous to Ontario, and the small Indian population is almost exclusively rural. Blacks first arrived as slaves to loyalist immigrants and, in the 19th century, as escaped slaves from the southern United States. Immigration continued during the 20th century, especially from Detroit, Mich. The black population is also small and is concentrated largely in Toronto, where most are probably of West Indian origin.

As well as immigration from abroad, Ontario has benefited from population movements within Canada. Approximately one-fourth of the net migration into Ontario came from other parts of the country. After Newfoundland joined the confederation in 1949, Newfoundlanders began to move to Ontario in increasing numbers, and there has been a continuing movement from Nova Scotia and New Brunswick. There are a few hundred thousand French-speaking Canadians living in Ontario, who were drawn from Quebec since the 19th century by the lumber industry and railroads of the north, the farms of the east, and the Cornwall industrial area. Ottawa contains the largest concentration of French Canadians, and there are sizable communities in Windsor, Toronto, and the Niagara Peninsula.

Emigration

Balanced against large-scale immigration is the fact that, as with Canada as a whole, there has long been substantial emigration from the province. As early as the 1850s, Ontarians were attracted by the westward movement in the United States, and outflow to the south has been important ever since. In the 20th century, much concern has been caused by what appears to be a disproportionate loss of professionally and technically skilled people to the United States. Similarly, many thousands of Ontarians have migrated to the Canadian West.

The result of the relationship between immigration and emigration is that natural increase has been the more important cause of population growth. The vital statistics (*i.e.*, the birth rate and death rate) and the rate of population growth for Ontario have been roughly the same as the Canadian rates for most of the 20th century. The provincial population more than doubled in the first half of the 20th century and doubled again during the next 30 years. The province is overwhelmingly urban, with 82 percent of its citizens living in cities and towns.

The economy. Ontario's strategic central location with respect to the Canadian provinces, its proximity to U.S. markets and coal supplies, its cheap power, its large and skilled labour force, its abundant natural resources and diversified transportation system, and its general attractiveness to both domestic and foreign investment, have made its economy the most productive in Canada. As with all mature economies, the bulk of employment and output is concentrated in the manufacturing and service industries, the latter having grown particularly fast in the later 20th century.

Federal economic policy remains the chief agent of control acting upon the private sector, but the provincial government has played an increasingly important role. In 1962 the province was divided into 10 economic regions for planning purposes, and the provincial government stimulated industrial research and development with various programs. Industrial disputes are regulated by the Ontario Labour Relations Board.

Agriculture. Ontario is one of the major agricultural regions of Canada. The most important cash crop is tobacco, but most farms are concerned with dairying or livestock. As a result of this specialization, corn (maize) acreage has expanded, though it is still necessary to import corn from the United States for livestock feed. The Niagara Peninsula and the Holland Marsh north of Toronto are the chief fruit- and vegetable-producing regions, while Kent and Essex counties in southern Ontario are the chief corn- and soybean-producing areas.

Thomas Kitchin/First Light

Pulp and paper mill at Thunder Bay, Ont.

Mining and forestry. Ontario is Canada's leading mining province. The province mines a large percentage of the world's nickel; this metal alone accounts for more than one-fourth of the total value of metal production. Copper production is the second most important metal, followed by uranium, zinc, gold, iron ore, platinum, and silver.

Although the forest products industry does not rank with agriculture and mining in terms of value of production, it is still one of the most important branches of the industry in Canada. Pulp and paper manufacture and fabricated wood products dominate the industry in Ontario. Commercial fishing was once concentrated in Lake Erie, but heavy pollution in that and other lakes sent the commercial fishing industry into decline. Sport fishing continues to be important for local residents and tourists. The main fish caught include yellow perch, pike, smelt, white bass, whitefish herring, and chub.

Sport fishing

Manufacturing. Ontario is the leading manufacturing province in Canada, employing almost half of the country's workers and generating the same proportion of the total value of production. Historically, this preeminence derived from the milling, farm implement, furniture, and textile industries of the 19th century. The Canadian protective tariff of 1878 encouraged a domestic steel industry at Hamilton and the location of many American branch plants in the province. Other leading industrial products include automobiles, textiles, processed foods, industrial machinery, electrical goods, farm implements, chemicals, rubber and synthetics, aircraft, and furniture.

Ontario and Quebec are Canada's leading provinces for petroleum refining. Conventional thermal power and hydropower together account for more than one-half of Ontario's electrical energy, and nuclear power accounts for the remainder.

Transportation. Bulk cargoes, chiefly consisting of mining and forestry products and prairie grains, are moved to the United States or overseas by the Great Lakes–St. Lawrence waterway system. Seagoing carriers bring imports from abroad by the same route. Toronto's port activity declined after the opening of the St. Lawrence Seaway.

The basic road pattern, laid out in the 1790s, is an east–west highway from the Quebec border to Windsor and a north–south expressway from Toronto. The Ontario section of the Trans-Canada Highway runs for 1,400 miles from Montreal through Ottawa to the Manitoba border. Capital and maintenance costs are high because the region's heavy snowfall and extreme temperature range make constant repairs necessary.

Ontario is crossed by two transcontinental lines and is bisected by one provincially owned railroad with its northern terminus at Moosonee on James Bay. Although there has been a reduction in passenger mileage in southern Ontario due to lack of revenue, this region's rail network—centred on Toronto—is Canada's most elaborate.

Toronto is also the focus for the province's air traffic. Its International Airport is now the country's largest and is a main centre of operations for both domestic and international flights. Other important airports are located at Ottawa, London, Sault Ste. Marie, and Thunder Bay.

Administration and social conditions. *Government.* The provincial government consists of the lieutenant governor, appointed by the governor-general in council, whose functions are purely formal; the Executive Council, or Cabinet; and the elected Legislative Assembly. The province is represented in the federal government by 24 senators and 99 elected members of the House of Commons. Manhood suffrage, except for Indians living on reservations, has been in effect since 1888. Women were given the vote in 1917, and in 1954 Indians were enfranchised.

Rural political domination

Since 1963, urban constituencies elect the majority in the legislature, but rural areas are still overrepresented. The Cabinet is headed by the premier (frequently called prime minister in Ontario) and tends to be dominated by rural interests. Northern Ontario is habitually underrepresented, as is the province's Roman Catholic population.

Ontario has three political parties—the Progressive Conservatives, the Liberals, and the New Democratic Party. The latter, founded in 1961, represents an amalgamation of the Co-operative Commonwealth Federation and sections of the trade-union movement. Stability and one-party dominance, however, have long been the mark of Ontario politics. The changing character of Ontario society has the potential to alter the tradition of one-party rule, but as long as three parties are competing for votes it seems an unlikely possibility.

Municipal government in Ontario consists of elected township and county councils in rural areas and of elected councils in cities and towns. Urban growth, the demand for enhanced services, and the movement of taxpayers to the suburbs promoted the federation of the city of Toronto and its 12 suburbs in 1954. The 13 components retained their existence as political entities, but certain general powers and responsibilities were transferred to a second tier of government. In 1965 the federation was adjusted, subsequently being composed of the city proper and five boroughs with roughly equal populations. These changes brought an efficient centralization of public services, enhanced the borrowing capacity of the municipality, and created a uniform tax system. Similar regional governments were instituted for Ottawa–Carleton in 1969 and for Niagara in 1970.

Education. Education costs have risen sharply because of increased enrollment in elementary and secondary schools and increased assistance to higher education. Undergraduate and graduate schools are rapidly expanding, and several universities, though legally autonomous, receive both operating and capital grants from the Department of University Affairs, which has a measure of control over university planning, intake, and curricula. The province has also established a number of community colleges and colleges of applied arts and technology to provide postsecondary technical education. Major universities, in addition to the University of Toronto, include Queen's University at Kingston, McMaster University at Hamilton, and Carleton University at Ottawa.

Health and welfare. In the field of social legislation, the province has entered into important shared-cost arrangements with the federal government. An Ontario Health Insurance Plan, for example, provides comprehensive medical insurance to all residents and is free for those with no taxable income.

Poverty among Indians, marginal farmers, and city dwellers is of continuing concern, but the proportion of the population on welfare has remained fairly constant at a relatively low level. Real wages have risen steadily and exceed the national average for manufacturing and other industrial personnel. The cost of living, however, is high compared to most of the other provinces, and substantial discrepancies exist between male and female wages, although legislation is now in force against some forms of such wage and job discrimination.

Cost of living

Cultural life. Toronto tends to dominate the province's cultural activities, although its preeminence is by no means exclusive. The first notable school of Canadian painting—that of the Group of Seven—was founded there in the 1920s. The capital city is the centre of Canada's English-language theatre, as well as the headquarters of national radio and television broadcasting. The city supports two symphony orchestras, two choirs, national opera and ballet companies, and many art galleries and museums.

Cultural institutions are also growing in other urban centres. Ottawa has the National Arts Centre, a symphony orchestra, museums, and art galleries. London is developing an artists' colony, while Stratford is already established as a fine dramatic arts centre and site of the annual Stratford Festival. The latter, which opened in 1953 in this pleasant western Ontario town, has had a substantial impact upon the cultural life of the province. There is also an annual George Bernard Shaw Festival at Niagara-on-the-Lake. Upper Canada Village near Morrisburg is a recreation of a 19th-century Ontario community, and Fort Henry at Kingston and Fort George at Niagara-on-the-Lake have been preserved.

HISTORY

The earliest known inhabitants of the Ontario region included the agricultural Huron, Tobacco, and Erie tribes of the south and the hunting groups of the Algonquin, Ojibwa, and Cree of the north. The French explorer Étienne Brûlé was the first known European to travel among them, doing so on an expedition to the Ottawa River in 1610–11. He was shortly followed by Samuel de Champlain and other French explorers, fur traders, and missionaries. The southern tribes were dispersed when the Iroquois destroyed the Jesuit mission at Fort Sainte-Marie in 1648–49, and France established Fort Frontenac (present-day Kingston) in 1673 to begin the military protection of its westward-spreading fur empire. When Canada was ceded to Great Britain in 1763, however, no French colonization had taken place, except for a small farming settlement in the Detroit area.

The Quebec Act of 1774 established Ontario as part of an extended colony ruled from Quebec. During the American Revolution the region was a base for loyalist and Indian attacks upon the American frontier, and in 1784 it was settled by approximately 10,000 loyalists and those of the Iroquois tribes who had fought for the British. The Constitutional Act of 1791 divided Quebec colony into Lower Canada, with a French majority, and the new loyalist province of Upper Canada. Upper Canada received representative government; provision was made for the support of the colonial administration and an established church by substantial land endowments called the Crown and Clergy Reserves.

Early European settlement

John Graves Simcoe, the vigorous first lieutenant governor of Upper Canada, supervised the introduction of English legal and local government practices, laid out the land-granting pattern, supported the construction of trunk roads, and fixed the capital at York (now Toronto). His policy of welcoming massive immigration from the United States was a source of tension between the newcomers and the established anti-U.S. loyalists, a rift that deepened during the War of 1812.

From 1815 to 1841 the province was dominated by a conservative coalition that was known as the "Family Compact" because it was alleged to be an elite tied together by family relationships. The coalition favoured the Anglican church and the Crown and Clergy Reserve system. Reformers demanded responsible government, and in 1837 a radical minority led by William Lyon Mackenzie attempted an unsuccessful revolt.

In 1841 the provinces of Upper and Lower Canada were united, and Upper Canada became known as Canada West. Responsible cabinet government was achieved with the formation of the Robert Baldwin–Louis Hippolyte Lafontaine ministry in 1848, and the present system of municipal government and the province's educational system were created. The 1850s brought the railways, the beginnings of industrialization, and the emergence of Toronto as a commercial rival for Montreal.

Political deadlock and the impetus of the new Ontario economy contributed to the movement for Canadian political union. Canadian federation—achieved in 1867—was brought about in large part by such Ontario politicians as the Conservative John A. Macdonald and the Liberal George Brown. Canada West became the province of Ontario, and the capital was located at Toronto. For a generation Ontario's government was headed by Oliver Mowat, the Liberal premier who won a boundary dispute with Manitoba that doubled the size of Ontario and confirmed the supremacy of provincial governments within their constitutionally assigned powers.

Location of the capital at Toronto

In the 20th century the chief concern of Ontario's governments has been the fostering of economic growth. The province has been transformed into a multicultural society engaged in a highly industrialized economy. With the harnessing of Niagara Falls in 1882, Ontario entered into an energy revolution that also encompasses the power potential of the north, the St. Lawrence River, and uranium-generated thermal power. Since 1914 there have been major discoveries of minerals, as well as expansion of the forest products industry. Tied to the emergence of the northern natural resources, the rise of manufacturing in southern Ontario was spurred by the two world wars to the production of a surplus for export. The provincial government has also been concerned with the growth of services for its burgeoning population. (S.F.W./Ed.)

Quebec

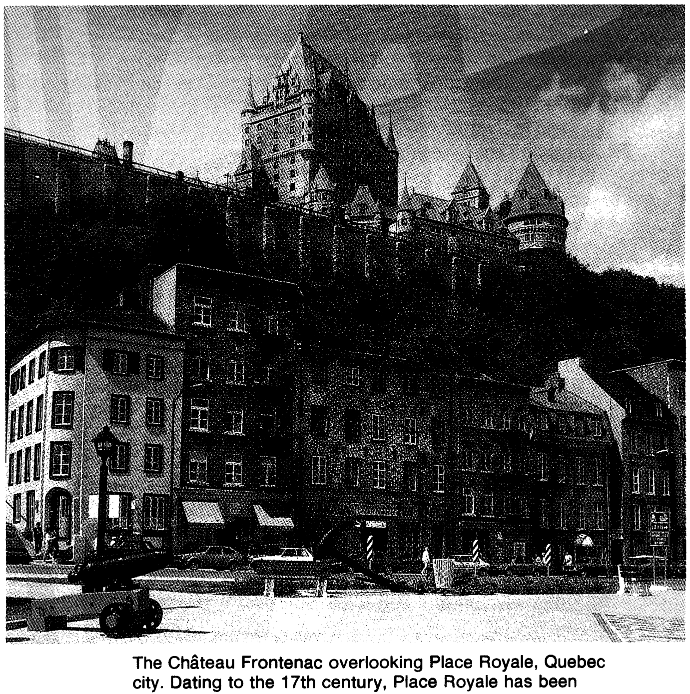
The Province of Quebec (French: Québec), with a total area of 594,860 square miles (1,540,680 square kilometres), is the largest Canadian province in size and is second only to Ontario in population. Its capital, Quebec city, is the oldest city of Canada; and its metropolis, Montreal, is the second largest city in Canada. It is bounded on the north by Hudson Strait and Ungava Bay, on the east by Labrador, on the southeast by the Gulf of St. Lawrence and New Brunswick, on the south by the United States (Maine, New Hampshire, Vermont, and New York), and on the west by Ontario and Hudson Bay.

To understand present-day Quebec, however, one must see the province against a background that goes back to the creation of the French colony in North America during the 16th century. Most observers would agree that the single most important theme in Quebec's history since the British acquisition of New France in 1763 has been the continuous attempt to achieve an accommodation between the numerically dominant French-speaking population and the economically dominant English-speaking one. Whatever changes in geographic size or political institutions have taken place in the province, life in Quebec has always been marked by a collective effort to maintain a distinct French-speaking society. This characteristic in the second half of the 20th century was at the core of debates over the federal structure of all Canada.

Creation of modern Quebec province

The present Province of Quebec was created in 1867, after being the colony of New France for more than two centuries until it was ceded to Britain in 1763. Named the Province of Quebec between 1763 and 1791, it then became the Province of Lower Canada until 1841, and then the District of Canada East until 1867. During these earlier periods its geographic boundaries were changed arbitrarily, and only in the 20th century, with the reacquisition of the northern part of Quebec, did it acquire its present size. Even today, however, there are problems about the eastern boundaries, because no Quebec government has accepted a 1927 decision of the British Privy Council to award Labrador to Newfoundland.

George Hunter



The Château Frontenac overlooking Place Royale, Quebec city. Dating to the 17th century, Place Royale has been largely restored to its appearance during the time of the French regime.

Quebec's size and its boundaries are not the most important influences on its life. The province was profoundly marked by 18th-century wars between France and Britain over their North American territories and by difficulties between the two linguistic groups since 1763, creating tensions that the social, economic, and political institutions of Canada and of Quebec have been unable to resolve. Because only in Quebec, New Brunswick, and at the level of the federal government is French an official language, French Canadians have felt that they are threatened as a minority group in Canada. In the past, control by the English minority in Quebec of most economic activities of the province had generally led to exclusion of the French Canadians from opportunities of economic advancement. In Quebec, where they constitute about 82 percent of the population, they maintain that the situation remains discriminatory.

Although people of goodwill on both sides have tried to find a lasting solution, the economic and social inequalities have created a growing nationalism among French Canadians and a feeling among some of them that only the separation of Quebec from Canada can solve their problems. Events since the late 1960s have shown, however, that neither extreme nationalism nor separatism is accepted by the majority of French speakers who live in Quebec, although there was a resurgence of nationalism in 1990.

PHYSICAL AND HUMAN GEOGRAPHY

The land. Quebec is a sort of "land's end," almost separated from the rest of Canada by Hudson Bay and Hudson Strait. Beginning in the 1960s, provincial maps issued by the Quebec government reflected its long-standing territorial dispute with the federal government by not showing the Labrador boundary, thus including the whole of the North Atlantic coast down to the Gulf of St. Lawrence as part of the province.

The Canadian Shield

Geographic regions. Quebec usually is divided into three major geographic areas that reflect its main geologic structure: the Canadian Shield (also called the Laurentian Shield), the populated plains of the St. Lawrence lowlands, and the Appalachians. The Canadian Shield covers about 80 percent of Quebec, comprising the entire region north of the St. Lawrence lowlands. Beginning at the foothills of the Laurentians (Les Laurentides), the oldest mountain range in the world, it runs northward to the Hudson Strait. The Shield is composed of three main subdivisions: the Laurentians, covered with trees, which have become a natural playground for summer and winter activities; the taiga, a region of stunted trees farther north; and the tundra, in which a continuously frozen ground, or permafrost, allows no trees to grow and where summer enlivens only reindeer moss and a few dwarf birches.

The fertile zone of Quebec and the most densely populated is the plains of the St. Lawrence lowlands, while the Appalachian region is composed of hills and plateaus, more or less undulating and rising to the higher mountain ranges in the United States. This region includes the Eastern Townships (the area south of the St. Lawrence lowlands) and the Gaspé Peninsula.

Climate has been a major factor in determining plant and animal life. The vegetation zone gradually moved northward after the glaciation period. North of the 56th parallel the treeless Arctic tundra is characterized by lichen, mosses, and peat bogs. In it live the polar bear, fox, and Arctic hare. Southward, between the 56th and 52nd parallels, is the taiga, with trees in sheltered areas and herds of caribou that sometimes number in the thousands. South of the 52nd parallel is a forest of spruce, fir, and pine, and in the St. Lawrence and the Ottawa river valleys are forests of maple, ash, beech, and oak. These forests and the rivers and lakes in and around them abound in animal life, including hundreds of bird species, more than 100 freshwater fish species, and numerous sea mammals in the St. Lawrence estuary.

Drainage. The landscape of Quebec is laced with thousands of lakes and rivers, which amount to a freshwater area of some 71,000 square miles. The St. Lawrence, which cuts across southern Quebec from west to east, is one of the world's greatest waterways. Among its tributary rivers are the Ottawa, the Saguenay, the Saint-Maurice, and the Manicouagan.

Climate. Quebec's climate is often extreme, sometimes severe. In Fort-Chimo, on Ungava Bay, the temperature ranges from an average of −11° F (−24° C) in January to 52° F (11° C) in July. In the south in Sherbrooke, it varies from a January average of 14° F (−10° C) to a July average of 68° F (20° C). Temperature changes may be as great as 30° F (17° C) in less than 24 hours. The period in which the snow remains on the ground differs from an average of 12 to 13 weeks on the Montreal plain to 23 weeks on the north coast of the St. Lawrence. The same variations exist for days without frost, Montreal having an average of 140 days without frost, the far north fewer than 80 days.

Patterns of human settlement

Settlement patterns. For all practical purposes, the inhabited part of Quebec is limited to the St. Lawrence lowlands and parts of the Appalachians and the Laurentians. More than 80 percent of its population lives within an area 200 miles long and 60 miles wide, one of the highest concentrations in Canada. About 80 percent of all Québécois live in towns, a very small number are classified as rural farmers, and the balance of the population is scattered in forestry, fishing, mining, and other types of activity in small settlements. The historical movement of the population in Quebec has been from large numbers of scattered, diversified settlements to an increasing concentration in urban areas. The shortage of fertile land suitable for cultivation prevented the development of a truly agricultural economy.

The rural settlements that existed during and after the period of French control were limited to the shores of the St. Lawrence, forming a continuous line between the urban centres of Montreal and Quebec city. Outside of Quebec city, Trois-Rivières, and Montreal, the land was divided into long, narrow, individual strips. A house was built at the end of each of these strips, on the side of the road that led to the towns, thereby forming a type of ribbon development. Rural Quebec had no villages until the end of the 18th century, and most rural parishes were created during the 19th. As each road became fully settled, a parallel road was opened farther inland, a process that was repeated until the whole of the St. Lawrence lowlands was occupied. This system of colonization, in which houses were located equidistantly along a road, allowed for maximum density in settlement, but it was not suitable for agricultural activities.

The very rapid urbanization of the province in the 20th century did not imply, therefore, the breakdown of a traditional peasant society. Although relatively isolated rural and fishing settlements existed on the north coast of the St. Lawrence and mining and foresting settlements in the Laurentians or in northern Quebec, these communities did not create any distinctive "folk" type of society that would have given a dominant cultural orientation to French Canada.

The future of Quebec was for a time linked with the development of the Canadian Shield, which contains one of the world's largest reserves of minerals. The first area to be developed in the 20th century was around the Noranda–Rouyn area, which is noted for copper, zinc, lead, gold, and silver. Continued demand for copper resulted in the opening up of new mines in Chibougamau and in Labrador, where large deposits of iron are located. Outside the shield, copper is mined at Murdochville in Gaspé, while the Eastern Townships supply nearly two-thirds of the Western world's output of asbestos.

Cheap waterpower, abundant raw materials, relatively low labour costs, and dense concentration of population in the St. Lawrence lowlands have been the main factors behind the development of industries in Quebec. The process began with the building of textile mills in the Eastern Townships, but industry gradually was concentrated in the Greater Montreal area.

Because of the present distribution of the population, Quebec can be thought of as divided between Greater Montreal, with more than two-fifths of the population, and the rest of the province. Although there are other fairly large cities in the province, such as Ottawa-Hull, Quebec, Chicoutimi-Jonquière, and Trois-Rivières, this division represents the basic structure of Quebec society. The dichotomy is most evident when outlying areas of Quebec are compared with Montreal. The Gaspé Peninsula is the poorest and least urbanized region; nearly one-fifth of its population earns its living through agriculture, and, in the summer months, tourism becomes the major source of income for the population.

Dichotomy of Quebec society

The people. Because the distribution of the various ethnic groups is the single most important factor in Quebec's social, economic, and political life, past and future trends must be taken into account. In the years of French colonization, from 1608 to 1760, only some 10,000 French came over to settle on the shores of the St. Lawrence. The approximately 60,000 persons who were "ceded" to Britain in 1763 were practically all Canadian-born, and most of them traced their descent through several generations of habitants, the name given to those born in New France.

The high birth rate among habitants explains the growth of French Canada. Between 1700 and 1760, the average yearly rate of birth was between 54 and 62 per 1,000 inhabitants, while the death rate was relatively low for the time: between 24 and 40 per thousand. After 1763, this extremely rapid natural growth of the population continued, and from 1800 to 1850 the total number of French Canadians increased from about 100,000 to 1,000,000.

The very favourable conditions for population growth ceased gradually to operate during the 19th century, however, for the fertile lands around the St. Lawrence were fully occupied, and increasing difficulties were encountered in maintaining satisfactory economic conditions for family life. French Canadians began a migration to the United States and other parts of Canada that between 1840 and 1940 is said to have taken one million persons from Quebec. Although efforts were made to open new rural areas in more northerly Quebec, the harsher conditions of life limited the new settlements. Trends were, in fact, away from rural settlements, and the new industries attracted an ever larger number into the towns. By the beginning of the 20th century, the birth rate had been reduced to about 38 per 1,000, while the death rate still hovered around 20 per 1,000. From then on, both figures declined steadily, especially the death rate. By the late 20th century, both the birth and death rates were slightly less than the national rates.

Other types of demographic changes have also been taking place. Demographers from the University of Montreal, for instance, have identified a long-term decline in the number of French-speaking Québécois, especially in the Montreal area. This trend is contributing to the gulf between Montreal and the rest of Quebec, with the result that most of the linguistic legislation prepared by the Quebec government is increasingly difficult to apply equitably to Montreal and to the rest of the province. Quebec society's rates of divorce, suicide, and illegitimate birth are among the highest in Canada.

American Indian population

The Indian population of Quebec has grown to more than 33,000, distributed among 39 different reserves and Indian communities. Some 5,400 Inuit live in northern Quebec. Inuit and Indian affairs in Canada are under federal jurisdiction, but, with the transfer of certain responsibilities to the province, a Direction Générale of northern Quebec has been created by the Quebec government to provide services for the Inuit population. Since establishment of this department a debate has developed between the federal and provincial governments over the responsibilities of each toward the Inuit.

Another distinctive characteristic of Quebec has been its religious homogeneity. During the period of New France, Roman Catholicism was the official religion. After 1760, freedom of religious practice was authorized by the British government. As the English population grew, so did Protestantism. Following the potato famine in Ireland in the mid-19th century that forced many Irish to migrate to Canada, the number of Roman Catholics in Quebec also increased. Many marriages took place between the Irish and the French Canadians, and today many persons in Quebec who have French as their mother tongue have Irish names. With the coming of such new immigrant groups as the Italians after World Wars I and II, the proportion of Roman Catholics again increased and at one time reached a peak of 88 percent. For a long period, religion was a basic mode of differentiation among the various groups; today this is not so, and it has ceased to be a factor in Quebec's political life. Other ethnic groups in Quebec include eastern Europeans, Portuguese, Greeks, Haitians, and Asians, particularly from Southeast Asia.

The economy. Until after the middle of the 20th century Quebec was considered to be largely an agricultural society, but the occupational concentration is now in industrial and service activities, indicating that Quebec has followed the general trend of industrialization in Canada. Within the major corporations, banks, or branches of multinational companies that operate in Quebec, a number of French-speaking Québécois have achieved high-ranking positions. Promotion, however, has been generally slow even in the nationally owned Canadian National Railway or Air Canada, both of which have their headquarters in Montreal. This problem has been an increasing point of social and political contention in the province.

Resources. Because of continuing difficulties in surveying the land, it is not yet known how extensive Quebec's natural resources are. Considerable mineral deposits are known to exist, however. Between 1910 and 1970, mineral production increased 100 times in gross dollar value (not accounting for inflation). This rather spectacular development was due to the discovery of extremely large iron-ore deposits in northern Quebec. Geologists estimate that these deposits are among the largest in the world. Two new towns were created in the north as a result, Schefferville and Gagnon; and a large port, Sept-Îles, was developed. Quebec also has reaped many economic benefits from the iron-ore fields of neighbouring western Labrador. Much of the ore extracted from the Labrador fields is shipped to nearby Quebec towns for processing. Other important minerals extracted in the province include gold, asbestos, clay, lime, sand and gravel, and stone.

Minerals, water-power, and forestry

Another of Quebec's major economic resources is electric power. With the nationalization of all electric power in Quebec in 1963, Hydro-Québec became the largest producer of electricity in Canada. Besides its more than 70 hydroelectric plants, Hydro-Québec has many thermal and nuclear plants. In 1967 the Research Institute in Electricity, the first organization of its kind in the world, was created near Montreal.

Forestry is the third great economic resource of Quebec. It is estimated that forests with economic potential cover about 295,000 square miles, more than three-fourths of which is exploited. Only about 42,000 square miles of forests are privately owned, accounting for about two-fifths of the annual timber cut. The remaining timberlands are public and under government management.

Agriculture. The concentration of population in the southern part of Quebec developed historically in response to the potential of the fine level soils of the lowlands and the undulating moraine-covered plateaus of the Appalachian region. The lowlands are used largely for animal feed, particularly hay, for dairy cattle. Tobacco is grown on the sandy soils of the Joliette area, vegetables in the area of muck soil south of Montreal, and apples on the slopes of the Monteregian Hills. The region of the Eastern Townships is another dairying centre, although only half the land in the area is cleared. Maple sugar and maple syrup provide additional income. The Gaspé Peninsula and the Canadian Shield region, with the exception of the western clay belt and the Lac-Saint-Jean Depression, provide only limited opportunities for agriculture.

Furs and fisheries. The fur trade has played a part in the economy of Quebec since the beginning of the French regime and is still important today in Ungava. Mink easily ranks first in production while other species include beaver, fox, lynx, otter, and chinchilla.

Maritime fisheries are important in the Gaspé Peninsula, the north shore, and the Magdalen Islands. The Quebec Department of Industry and Commerce operates a network of cold-storage plants. Principal species include cod, herring, redfish, lobster, and salmon.

Regression in the economy

Industry. A number of unfavourable factors affect Quebec's economy, and such areas as textiles have definitely regressed. Slowdowns have been recurrent in the pulp and paper, food and beverages, and construction industries and in other related services. The per capita income has also lagged significantly behind that of Ontario, the other highly industrialized province. A significant proportion of Canada's organic and inorganic heavy chemicals is produced in Quebec. The province is a leader in petroleum refining, with major production in the Montreal region, and Quebec's furniture industry is the oldest in Canada. The industrial core of the province is Greater Montreal, where some 7,000 plants account for most of Quebec's manufacturing output. Montreal is a leading producer of aircraft, railway locomotives, leather footwear, and chemical and pharmaceutical products; other industries include flour milling, tobacco, brewing, and distillery.

Transportation. Quebec is fully integrated in the general transportation system of Canada and of North America. Except in northern Quebec, neither climate nor distance are serious problems. The major handicap in the development of transportation is the low population density of the regions outside the urban area of the St. Lawrence valley and the restricted flow of goods and persons to the north.

The transportation system is largely oriented on the basis of the geographic position of Montreal, the major crossroads for moving persons and goods in and out of Quebec, whether by road, water, or air. Montreal has two international airports, one in Dorval and another in Mirabel. Quebec city also accommodates international passengers. The internal air traffic for Quebec province is provided by two companies. Montreal is both a major ocean port and, by virtue of the St. Lawrence Seaway, which allows ships to travel a total of 2,300 miles from the North Atlantic to the ports of the Great Lakes, a major inland port as well.

The railway system of Quebec is practically restricted to the St. Lawrence Plain, with a few branch lines of the two major Canadian companies. Three privately owned railways transport iron ore from northern Quebec.

Administration and social conditions. *Government.* Quebec's administrative system can be defined as government through parliamentary democracy. An elected unicameral National Assembly is the equivalent of the parliamentary institutions of other Canadian provinces; its second chamber was abolished in 1968. A lieutenant governor represents the British monarch. The Executive Council, or Cabinet, is headed by a prime minister who is responsible to the National Assembly for all legislation within provincial jurisdiction. The Executive Council has responsibility for preparing legislation for the National Assembly. The administration follows the parliamentary principle that a prime minister and his Cabinet will remain in power as long as the prime minister is able to command a majority in parliament.

Branches of government

An unusual characteristic of Quebec is its administration of justice. Although its Department of Justice, as in other Canadian provinces, has a dual responsibility for criminal and civil laws, the civil law of Quebec is different from that in the other Canadian provinces. It follows the civil code of the French rather than the English common law followed by the rest of Canada. Quebec also has its own provincial police, La Sûreté Provinciale du Québec, who have taken over this responsibility from the federal Royal Canadian Mounted Police. Although the economy is shared with an English-speaking minority, political life has for some time been dominated by French-speaking Canadians. From 1936 to 1976 the parliamentary government was in the hands of either Union Nationale or the Liberal Party, both controlled by politicians of French descent. In 1976 a new separatist-minded party, Parti Québécois, won control and, until they were ousted in 1985, fought a continuing battle against its Anglophone citizens and against the federal power in Ottawa.

Education. Nothing shows more clearly the complex nature of Quebec than its educational system. Organized originally along religious and linguistic lines and largely privately financed, it has become, for all practical purposes, a public system since 1964 under the provincial Department of Education. Larger sums are spent each year on education. Although the religious dichotomy between Protestants and Roman Catholics has been maintained, it is largely along linguistic lines that the separation continues. A problem causing tensions is the growth in the number of English-speaking pupils, while, because of the low birth rate, the number of French-speaking pupils at the elementary level is decreasing. The law permits parents to choose the language of education, and increasing numbers of persons are choosing English.

Higher education

Organized on the normal structure of preschool, elementary, secondary, and college–university levels, the educational system has created the Colleges of General and Professional Studies, which act as two-year pre-university-level institutions. There are three English-speaking and four French-speaking universities; of the latter, the University of Quebec has numerous remote campuses, including those at Montreal, Trois-Rivières, and Chicoutimi.

Health and welfare. The provincial government has an integrated system of health and social services. More than 200 hospitals are under government administration to serve the health care needs of the people. Social assistance is available for the indigent, the elderly, dependent children, and others in need. Although sharing with the rest of Canada one of the world's highest standards of living, Quebec has a larger proportion of unemployed and persons on social security than the Canadian average. The cost of social aid is one of the highest among provincial governments in Canada. This situation is explained by the fact that age, health, level of education and training, and similar factors place a higher percentage of the population below the requirements of a modern industrialized society. In the late 20th century the Quebec government developed minimum-wage legislation, reorganized its health service, and introduced universal medicare.

The relation among economic development, social conditions, and educational level remains one of the basic causes of the political tension between French and English, on the one hand, and between the federal and the provincial governments, on the other. Although Quebec has transformed some of its institutions in order to meet the difficult demands of an industrial society, the changes are yet too recent to be able to determine whether they will improve the present situation permanently.

Cultural life. In many ways, Quebec is a smaller plural society within the larger pluralism of Canada: that is to say, it is nearly as difficult to define the cultural identity of Quebec society as it is to define that of Canada as a whole. Although a minimum of overall linguistic identity does exist in French Canadian life in Quebec, there are also many cultural differences between its French-speaking groups in east Montreal and similar groups in such smaller centres as Abitibi, Lac-Saint-Jean, Gaspé, or the Eastern Townships. Regional variations heightened by immigration have produced a sociocultural fragmentation that in part explains the differences in political votes, religious behaviour, and even the quality in the use of the French language. Furthermore, there are class–cultural differences and a strong elitist tradition explaining the high social status of such professions as medicine, law, the clergy, and business management. Even the new middle class of French Canada, which has appeared since the development of urban industrial society, is more politically aggressive in Montreal than in Quebec city, for the obvious reason that in Montreal their socioeconomic status is most difficult. Because of the cultural variations within Quebec, it is often difficult and sometimes impossible to obtain unanimity in political decisions that touch cultural or educational questions.

Debate over Quebec's cultural identity

While claims of a unique French-speaking culture, based on certain traditions brought from France, go back to the 19th century, a debate exists as to the exact nature of Quebec's cultural identity beyond the simple fact of dominant use of the French language. In many ways cultural values have been called up in an attempt to create a distinctive society. This is reflected in the government's fostering of a "Quebec" national culture, which was largely nonexistent when French-speaking Québécois used to call themselves French Canadians. Cultural values have thus been a sign of political identity as much as, or more than, the expression of long tradition.

To ensure the continuation of this Québécois identity, the Quebec government has created a number of institutions aimed at fostering cultural life. Foremost among these institutions is the Ministry of Cultural Affairs, which is responsible for improving the quality of the language used and for stimulating cultural, literary, and other artistic activities. Created in 1961, it was the first of its kind in North America. It not only gives direct financial aid to such state cultural bodies as museums and helps more than 60 theatrical, ballet, and musical companies, but it also has contributed for a number of years to book publishing and to public libraries for book buying. One of the major responsibilities of the department is to develop cultural links with other French-speaking countries. Another of its activities has been the development of regional cultural centres within Quebec to foster regional cultural life. Another institution created by the Quebec government in its efforts to stimulate French-language cultural development is Radio-Québec, founded in 1968 to develop cultural and educational programs. In 1969 its scope was broadened to take in television as well, and it was renamed Quebec Broadcasting Bureau. Radio-Canada and the TVA Network also broadcast in French as well as English. There are French- and English-language daily newspapers, the French-language papers being more numerous by far. The largest dailies of both languages are published in Montreal.

Besides the official cultural institutions, Quebec possesses an extremely large number of private artistic organizations, ranging from theatre companies to filmmaking. The most spectacular cultural development in Quebec, however, has been that of the *chansonniers,* who represent a cross between poets and songwriters. Their popularity, especially among the younger generation, arises from the fact that their songs reflect the present search for cultural and political identity. The *chansonniers* are involved frequently in political activities and are identified largely with the nationalist movement. Music and painting also share in this artistic revival, as does literature.

HISTORY

When in 1534 the French explorer Jacques Cartier landed at present-day Gaspé and took possession of the land in the name of the king of France, he brought with him the traditions of mercantile expansion of 16th-century Europe to this land where Indians and Inuit had been living for some thousands of years. There is a debate among historians, however, as to when the real history of Quebec should begin. Because the Province of Quebec as a political and geographic entity was created by the proclamation of 1763, the notion is sometimes also advanced that its real history should start with the capitulation of the French army in 1760, although Quebec city was founded in 1608.

Problems of political identity

The various definitions given by historians are not simply semantic questions, for they contain diverse assumptions concerning the political identity of the Quebec government. For example, there is a political tradition among French Canadians that the government of Quebec is also the government of the French Canadian people, and, therefore, they are heir to what was New France. The 1966–67 *Annuaire du Québec* (*Québec Yearbook*) states this claim most clearly:

> Quebec is a state with limited responsibilities that belongs to the Canadian Federation as a province. It is also the national state of the French-Canadians and exercises its governmental prerogatives, in the areas of its responsibilities, on the majority of the heirs of those who colonized New France.

To this, some authorities have replied that the territories covered by New France and those now included in the Province of Quebec cannot be equated. Although New France began with the founding of three cities—Quebec city in 1608, Trois-Rivières in 1616, and Montreal in 1642—it finally included territories that extended west in what is now the United States to the Ohio and Mississippi rivers. Even if the British government, by the Quebec Act of 1774, did in fact include practically all the territories of

New France in the new Province of Quebec, this situation lasted only briefly.

Some English-speaking historians assert that the Quebec Act created what is now Quebec as well as the practice of trying to fuse British and French institutions in the new political entity. The new British colony was, thus, to be administered by a governor and a council, using British criminal law and French civil law. Whatever the British government intended, however, when the composition of Quebec's population gradually changed as a result of increasing English-speaking immigration, it became increasingly difficult to carry out a policy that could give satisfaction to both the English- and French-speaking groups. In 1775, the year the American Revolution broke out, Quebec city was besieged by American troops, and Montreal was occupied. When peace was restored in 1783, the loyalists who had fled from the United States were settled west of Ottawa River, in what became the province of Ontario. This was the beginning of the basic geographic dichotomy in Canada between French and English. In 1791, Canada was split into Lower Canada (Quebec) and Upper Canada (the future Ontario).

Beginnings of ethnic inequities

Furthermore, although throughout the province the rural population remained overwhelmingly French, Montreal became the domain of the English merchants, who were bitterly anti-French. The metropolis of Canada was to have an English-speaking majority until the middle of the 19th century, and, even after that, it took nearly a century before French speakers gained control of its economic life. Discrimination existed between the two groups not only in economic, political, and religious activities but also in such other fields as education. Gradually, two different educational systems came into being. English-speaking McGill University was opened in 1821, but it was not until 1852 that the French-speaking Quebec Seminary, founded in 1668, became Laval University.

During the first part of the 19th century, the causes for conflict between the two groups increased with the rapid growth of the English-speaking population in Canada. The English merchants of Montreal tried in 1822 to obtain an Act of Union that would have united Lower and Upper Canada and given them an English-speaking majority in the country as a whole. The reaction against this attempt among French Canadians was strong and prepared the way for the 1837 rebellion. This rebellion, the first major manifestation of political nationalism among French Canadians, was led by Louis-Joseph Papineau, whose Patriote Party became a centre for radical politics. After the rebellion was put down, the British government sent out the Earl of Durham to investigate; his report, published in 1839, offended French Canadians by referring to them as a people without a history or culture and by characterizing the situation in Lower Canada as "a war between two races." The report also suggested the setting up of responsible government in Canada as a solution to the tensions between the two groups. In 1841 a new Act of Union joined the provinces of Upper and Lower Canada, and, in 1867, the British North American Act created the confederation of Canada by the federation of the four provinces of Nova Scotia, New Brunswick, Quebec, and Ontario.

From then on, French Canadian nationalism became a permanent feature of Canadian as well as Quebec politics. Doctrines of papal supremacy over national authority introduced the idea of the religious mission of French Canadians in North America. Under the leadership of such men as Henri Bourassa and the abbé Lionel Groulx, the province evolved its special vocation as the "political home" of French Canadians, and the government of that province assumed special responsibility for the defense of French culture. This situation also resulted in the doctrine of provincial autonomy that was used by Prime Minister Maurice Duplessis between 1936 and 1957.

Revival of nationalism and separatism

French Canadian nationalism also led to the "quiet revolution" of the Liberal government under Jean Lesage, who took office in 1960, and to the not-so-quiet revolution of a terrorist group known as the Front de Libération du Québec (FLQ), which was responsible for sporadic violence and the murder in 1970 of Quebec's labour minister, Pierre Laporte. The creation of the Parti Québécois in 1970 brought into being a new forum of Quebec nationalism, one that is no longer strictly French Canadian: it has English-speaking members as well as members of other ethnic groups, and its advocacy of separation from Canada is based on issues of economic and social development. In 1976 the party won the general elections and control of the provincial parliament, under the premiership of René Lévesque. In a referendum held in May 1980 the Quebec electorate rejected the opportunity to negotiate with the national government for sovereignty-association status (see above *Canada since 1920*). In elections held the following year the Parti Québécois maintained its majority in the provincial parliament, but in 1985 it was defeated and a Liberal Party government was installed. From that time the appeal of the political ideology of separatism declined steadily, although some revival occurred in 1990 when the Meech Lake accords, which would have recognized Quebec as a distinct society, failed to be ratified by the Canadian Parliament. (P.Ga.)

PRAIRIE PROVINCES

Alberta

Alberta, sparsely populated but possessing rich resources and beautiful scenery, is the most westerly of the three Prairie Provinces occupying the continental heartland of western Canada. With a land and freshwater area of about 255,285 square miles (661,190 square kilometres), it remains one of the most attractive of the underpopulated areas of the globe.

Location and general character

Following the earliest explorations of fur traders, the settlement of its plains and parklands led to the rapid development of agriculture. Subsequent exploitation of rich oil, gas, coal, and timber resources has led to further population growth, with steadily increasing urbanization and industrialization. The scenery of the mountain parks in the west is internationally renowned.

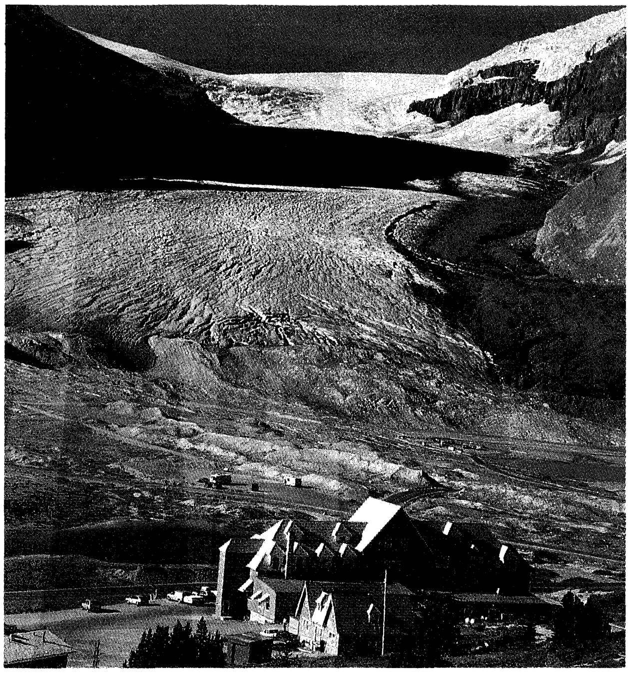
Relative isolation from the more populous eastern regions of the nation has slowed the development of industries needing mass markets. However, natural routes to the north have made Alberta the major jumping-off point for development of the Arctic.

Alberta is 756 miles (1,216 kilometres) in extent from north to south, and 404 miles (650 kilometres) across at greatest width. To the east, the 110th meridian is the boundary with its prairie neighbour Saskatchewan. Its western boundary with British Columbia is formed by the 120th meridian and the crest of the Rocky Mountains. Northward, beyond the 60th parallel boundary the adjoining Northwest Territories slopes to the Arctic Ocean, while the 49th parallel in the south forms the international boundary between Alberta and the state of Montana in the United States. Freshwater surfaces cover about 6,485 square miles (16,800 square kilometres) of the province. Alberta was established as a district of the Northwest Territories in 1882 and was enlarged to its present boundaries on becoming a province in 1905. The provincial government has its seat in Edmonton.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief and drainage.* Along the southwest border, Mount Columbia (12,294 feet [3,747 metres]) is the highest of 26 magnificent peaks reaching above 11,000 feet. Beyond the narrow foothills, soft rocks form an undulating plateau falling from over 3,000 feet to under 1,000 feet in the northeast, where the ancient rocks of the Precambrian shield outcrop. Outliers of higher ground include the Cypress and Swan hills and the Caribou Mountains.

The Columbia Icefield is the source of two major Alberta rivers, the Athabasca and the North Saskatchewan. The first flows northeast to the Lake Athabasca–Lake Claire



The Athabasca Glacier on the Columbia Icefield in the Rocky Mountains, Jasper National Park, Alta.

Alec Pytlowany/Masterfile

complex, where it is joined by the Peace River. The Slave River drains the area north toward the Arctic Ocean, and the lowest point in the province (573 feet) occurs in its valley.

Southern drainage system

Apart from a small area drained by the Milk River into the Missouri, the south of the province is dominated by the Saskatchewan River system. The Oldman, Bow, and Red Deer rivers combine to form the South Saskatchewan, which flows east with the North Saskatchewan toward Hudson Bay. Many of the rivers flow through deeply incised valleys, in which the eroded, barren landscapes known as badlands may develop. Those of the Red Deer are famous for their rich deposits of dinosaur remains.

Climate. Alberta has a continental climate, with more sunshine than any other Canadian province. Winters are dry, sunny, and cold, though in the south the Chinook winds, which comprise warm, dry airflows of Pacific origin descending the western slopes of the Rockies, may raise temperatures by 40° F (22° C) in 10 minutes. Summers are warm and wetter (except in drought years), with occasional destructive hailstorms and tornadoes, the latter having increased in frequency in the 1980s. Edmonton's mean and extreme temperatures are, in January, 5° F and −57° F (−15° C and −49° C), while in July they are 63° F and 95° F (17° C and 35° C). Annual precipitation averages 18 inches (466 millimetres), about half of it falling between June and August.

Plant and animal life. Chernozems—rich brown and black soils with a deep humus layer—underlie the prairies and parklands. Blue grama and other grasses dominate the prairie area in the south, although cacti, tumbleweeds, and sagebrush are more conspicuous; few trees grow naturally outside river valleys. The transitional parklands have aspen bluffs, increasing in size until forest dominates. The less fertile gray-wooded soils known as podzols underlie the extensive mixed and coniferous forests of the foothills and the north, where various combinations of aspen, white spruce, jack and lodgepole pines, and balsam fir are found. Black spruce and tamarack grow over accumulations of organic peat in extensive bogs. Engelmann spruce and subalpine fir are important trees in the mountain forests. The wild rose, Alberta's provincial flower, is widespread.

Meadowlarks and pronghorn are conspicuous on the prairies, and badlands support rattlesnakes, horned lizards, and scorpions. White-tailed deer and coyotes are familiar in parkland areas, while the great horned owl, the provincial bird, rears its young in vacated crows' nests. The northern forests house a wealth of fur bearers and big game, including the moose, wapiti, caribou, and black bear. Gray jays visit campsites, and loons are heard on the many lakes. Mountain goats, bighorn sheep, and grizzly bears, together with Clark's nutcrackers and golden eagles, are characteristic of the mountains, while rivers and lakes throughout the province support trout, whitefish, and pike; ducks nest by every slough. Examples of most habitats and some threatened species receive partial protection in national and provincial parks, wilderness areas, and ecological reserves. The endangered peregrine falcon, however, also flys among the high-rise buildings of Edmonton and Calgary.

Threatened habitats

Settlement patterns. The agricultural southeast is fairly uniform, though with considerable variations in landscape, farming practice, and density of settlement, while the isolated Peace River block forms a basically agricultural area to the northwest. The broad strip of forested foothills is much visited for hunting and other recreational purposes, utilizing the network of forestry roads. Some parts of the northern forested area are little changed from fur-trading days, with a scattered population of trappers and fishermen. The road network is advancing, however, and the province is committed to the development of pulp mills there. The mountain national parks each have small permanent populations and many visitors.

The square pattern of the Dominion land survey dominates much of the rural landscape. The survey established townships with 36 sections, each one mile square. From the air, the resulting great square fields can be seen spreading from horizon to horizon, while on the ground the straight roads seem to stretch endlessly.

Homesteading introduced some individuality into this pattern: log huts, elegant frame buildings, and modern houses were often occupied successively on the same farm and may still survive, although rural depopulation has caused abandonment of many dwellings. A network of dugouts and irrigation canals provides water in dry areas. Also, Indian reserves, presurvey settlements, and the communal farms of some religious groups make their own distinctive contributions to the landscape.

Groups of highly visible grain elevators, set at intervals along the railways, are characteristic sights in the province. Most villages have grown up around such a group, often only on one side of the railway track. A grid street plan and frame houses are typical, along with small shops, a hotel, and several churches.

Most urban centres gain character from their often splendid settings rather than from outstanding buildings. Such settlements as Peace River, in a spectacular river valley, and Drumheller, among extensive badlands, are noteworthy but not untypical. Most towns have a basically rectangular grid pattern, perhaps elongated along river frontage or a railway.

Each of Alberta's major cities serves a large rural area: Lethbridge and Medicine Hat, the southwest and southeast, respectively; Grande Prairie, the Peace River country; and Red Deer, the central region. Edmonton and Calgary have a much wider influence. Each is developing the high-rise buildings and expressways characteristic of cities in more populated centres, while retaining enough parkland to remain an attractive place to live.

Varied ethnicity

The people. In terms of national and cultural origins, Alberta is one of the most varied of the Canadian provinces. The early settlement around the turn of the century, as well as immigration following World War II, contributed to this diversity. Persons of British origin or descent are the most numerous, followed by Germans, Ukrainians, Scandinavians, French, and Dutch. The Indian population constitutes about 3 percent, and there are also Polish, Austrian, Russian, Chinese, Japanese, Lebanese, Vietnamese, and other ethnic elements. There are about 20 religious denominations, with the Roman Catholic, the United Church of Canada, and the Anglican Church of Canada predominating. Other groups include Mormons, Mennonites, Hutterites, and Muslims.

This human mosaic has provided a valuable cultural legacy for Alberta. This has found expression in festive occasions, costumes, the performing and fine arts, and in handicrafts. In domestic and religious architecture, this heritage is evident in areas of ethnic concentration such as the regions north and east of Edmonton; the Nikka Yuko centennial garden in Lethbridge; and the Hutterite settlements around the province.

The most densely populated area of Alberta is the central and southwestern portion, east of the foothills. The lowest density, averaging about one person per square mile, is in the northern portion of Alberta, the foothills and the mountain regions, which, together, occupy about three-quarters of the province. Alberta's population growth leveled off because of the economic slowdown of the 1980s, but rural depopulation continues.

The economy. *Resources.* Alberta holds the bulk of Canada's known fossil fuels. Oil and gas occur widely, and major deposits of heavy crude oil are being exploited in the Lloydminster, Cold Lake, and Fort McMurray regions. Alberta produces about 90 percent of Canada's natural gas and crude oil and a significant proportion of its coal. Fluctuations in world oil prices seriously affect the economy. Other mineral resources include sands, gravels, bentonite, clay, limestone, salt, phosphates, and sodium sulphate; sulphur and helium are produced along with the natural gas.

Forests cover more than one-half of the surface of the province. Aspen, birch, white spruce, and pines are the principal commercial species and are exploited for lumber, plywood, and pulp.

Industrial ecology problems

Most of the province has an abundance of lakes and rivers, a considerable number of which are exploited for irrigation and hydroelectric power. The ecological damage caused by these developments, weighed against the economic gains, has caused widespread controversy. Also significant as a resource are Alberta's wildlife and spectacular scenery, which, in combination with historic cultural attractions, form the basis of an expanding tourist industry.

Agriculture, forestry, and fishing. Although fisheries and furs remain important in the undeveloped areas of the province, several million acres (some irrigated) are under crops, mostly wheat, followed by barley, hay, and oats. There is increasing diversification, with rapeseed (canola), flaxseed, sugar beets, potatoes, rye, and mustard becoming important. Cattle, including dairy cattle, are also significant, together with swine, sheep, and poultry.

Industry. Natural resources are largely developed by private industry, controlled by the Canadian and provincial governments through legislation and regulations. Much of the industry is managed and financed outside Canada. Extensive government research aids resource development, and practical assistance is provided to agriculture.

Major manufacturing industries are foods and beverages, petroleum and chemical products, metals, transport equipment, and wood and paper. During the 1980s there also was considerable development of high-technology industries in such fields as computers, telecommunications, and medical equipment. Construction is linked to the development of gas and oil resources but also includes a proportion of residential building.

Transportation. Southern Alberta communications are dominated by east–west routes that thread the few mountain passes and spread over the plains. There are good communications south into the United States, and from Edmonton air and surface routes fan out to serve the Arctic from Alaska to the northern Canadian islands. The Athabasca and Slave rivers are active in the carrying of freight to the Arctic. The east–west Canadian Pacific and Canadian National railways have a network of branch lines serving agricultural and northern areas. Oil and gas pipelines link Alberta to markets as far away as Quebec and California. Alberta's major highways include the Trans-Canada and Yellowhead east–west routes and the Mackenzie and Alaska highways to the north and northwest.

Air transport has been used in development of northern resources since 1921; in the late 20th century there were elaborate air charter services as well as scheduled regional, national, and international services. Edmonton and Calgary have international airports.

Administration and social conditions. *Government and politics.* The Canadian Constitution and the Alberta Act (1905) provide the constitutional framework for the province. In line with other Canadian provinces, Alberta has a lieutenant governor (functioning as representative of the British monarch), an elected legislative assembly, and an executive council. The lieutenant governor is appointed by the federal government for a five-year term. The premier heads the executive council, which is responsible for administering laws and appropriations approved by the legislature. From 1935 to 1971 the Social Credit Party was dominant. The Progressive Conservative Party thereafter revived to lead the province, with Liberals and New Democrats forming a large opposition.

The Mounties

There is no provincial police force. The Royal Canadian Mounted Police is under contract to the province and to some municipalities, while in others there are municipal police forces. Correctional and rehabilitation services are provided under federal and provincial jurisdiction.

Municipal administration takes a variety of forms, appearing as cities, towns, villages, counties, municipal districts, and improvement districts (special areas). Mayors, councils, and school and hospital boards are elected, while the major sources of municipal revenue are the property tax, the local-improvement tax, and the business tax. Substantial grants are nevertheless made by the provincial authorities, although some municipalities derive additional revenue from the operation of utilities.

Education. Education is on a 12-grade system and includes a large but separate Roman Catholic school system, tax-supported. The province outlines the school curriculum and maintains general supervision of the elementary and secondary school systems, but administration is through elected school boards. The School Foundation Program Fund, pooled from provincial grants and local property taxes, provides funds—distributed back to the school boards—for the school systems.

Universities are located in Edmonton, Calgary, Lethbridge, and Athabasca, the latter specializing in distance education, a system of off-campus learning. Institutes of technology in Calgary, Edmonton, and Stony Plain provide extensive vocational training. A network of community and agricultural colleges supplement the system. Financial assistance for students is obtainable through the Alberta Students Assistance Board.

Health and welfare. The Alberta health insurance plan and the Alberta hospitalization benefit plan provide for health services through payment of an annual premium but with subsidies for low-income families. The Alberta Blue Cross plan and private insurance plans provide for certain supplementary hospital, drug, and other services, while mental hospitals are provided under a special provincial act. The costs of social assistance programs are borne jointly by federal, provincial, and municipal authorities.

Cultural life. Although the province is young, interest in preserving its heritage and fine arts is strong. The Cultural Development Branch of the provincial government gives practical aid and encouragement in all fields through such programs as performing-arts tours and exhibitions, leadership training programs and workshops in all phases of performing and expressive arts, consultative services, and financial assistance.

Folk tradition

The arts. The early artistic history of the province is largely one of rich oral folklore, although a major Icelandic poet, Stephan G. Stephansson, lived near Markerville, and his house is now a historic site. Since the 1940s a limited number of noteworthy artists (all immigrants to the province) have emerged, including the painter Illingworth Kerr and such writers as W.O. Mitchell and Rudy Wiebe. Interest in cultural heritage and the arts has increased rapidly since the 1960s, and Calgary, Edmonton, and Banff exhibit a rich and diverse artistic life with thriving communities of artists in all fields, many of national and some of international recognition.

Cultural institutions. A network of museums and historic parks now interpret the natural, historic, and artistic heritage of the province. The Provincial Museum of Alberta was opened in Edmonton in 1967. The Glenbow-Alberta Institute in Calgary has important collections of history and art, and the Tyrrell Museum of Palaeontology in Drumheller is the world's largest fossil museum, with particularly fine displays of dinosaurs. The latter operates in association with Dinosaur Provincial Park near Brooks, the first World Heritage Site to be designated for its fossil resources. Head-Smashed-In Buffalo Jump, another World Heritage Site near Fort MacLeod, features a buffalo jump that was used for 6,000 years. Other notable institutions include the Ukrainian Cultural Heritage Village east of Edmonton, the Crowsnest Ecomuseum, Calgary Zoo, and Edmonton's Space Sciences Centre and Art Gallery.

The performing arts are served by the Jubilee auditoriums in Edmonton and Calgary and the Calgary Centre for the Performing Arts, which are used by local symphony orchestras and opera and ballet companies, and by the Citadel Theatre complex in Edmonton. A lively performance scene exists in less formal settings, with major folk and jazz festivals, Edmonton's Fringe Theatre Festival, and a host of theatre and other groups performing outdoors, and in a variety of small theatres and adapted buildings. Smaller centres are served by provincial educational television and radio. Support for many of the arts is provided by the internationally renowned Banff Centre for Continuing Education, the Provincial Department of Culture and Multiculturalism, and a number of independent funding agencies.

Recreation. Alberta has the largest area devoted to national parks of any province in Canada. The national parks and several provincial parks are heavily utilized for such recreational activities as downhill and cross-country skiing, mountaineering, camping, hiking, and birdwatching. Ethnic and Indian arts and crafts are featured at such annual events as the Banff Indian Days, the Calgary Stampede, the Edmonton Heritage Festival, and many local fairs and rodeos. Ice skating and curling are important sports during the long winters, and professional ice hockey and football are extremely popular. The 1988 Winter Olympics, held in Calgary, centred world attention on the city and the province. (R.O.H./D.A.E.S.)

Traditional cultural and sporting events

HISTORY

The area now known as Alberta has been inhabited by various Indian groups for at least 10,000 years. European explorers first appeared in the 1750s as the fur trade expanded across western North America. Two rivals, the Hudson's Bay Company and the North West Company, began building trading posts in the last quarter of the 18th century along the major northern rivers, most notably the Athabasca, North Saskatchewan, and Peace. From 1821, when the companies merged, until 1870, when this region was transferred to the Dominion of Canada, the Hudson's Bay Company controlled and governed the area, which was populated by Indians, the métis, and a few European fur traders, missionaries, and settlers.

After 1870, settlement in southern Alberta began, based on a ranching economy. The Indian tribes had been decimated by European diseases and the disappearance of the buffalo, their main source of livelihood. The signing of treaties relegated the remaining Indians to reservations, but not before the abuses of unscrupulous whiskey traders had hastened the creation of the North West Mounted Police. In 1874 they established Fort Macleod and laid the bases of Canadian law enforcement in Alberta.

The Dominion Lands Act of 1872, which provided low-cost homesteads, the building of the Canadian Pacific Railway, which reached Calgary in 1883, and vigorous promotional campaigns brought an influx of settlers from eastern Canada, the United States, and Europe. By 1901 the population had reached 73,000, and by 1911 it had ballooned to 374,000. The development of earlier-maturing and more disease-resistant varieties of wheat made crop farming less risky and, in northern areas, newly feasible. Subsequently a wheat-based economy expanded throughout most of the arable parts of the province.

Responsible government developed gradually from 1875, when the North West Territories Act went into effect, until 1897, when a fully responsible legislative assembly was elected. Made a district of the Northwest Territories in 1882, Alberta was enlarged to its present boundaries in 1905 when it was made a province of Canada, although crown lands and natural resources remained under federal control until 1930. Edmonton, as the capital and distribution centre, grew rapidly, as did other urban centres. Calgary boomed with the discovery of oil at Turner Valley in 1914. Medicine Hat and Lethbridge, the latter a coal-mining area since 1869, developed into large distribution centres. Railways spread over most of the province, increasing agricultural development and providing a ready market for the province's vast coal deposits.

Aspects of economic growth and instability

During World War I the growth of population and the economy slowed. A postwar slump in wheat prices was a major factor leading to agrarian political discontent and the election success of the United Farmers Party of Alberta in 1921. During the 1920s the economy improved and population again increased, but the depression of the 1930s had a devastating effect. In 1935 the Social Credit Party, with a new monetary policy, was elected and retained power for 36 years. Not until World War II did the economy begin to recover. The discovery of oil at Leduc in 1947, as well as discoveries of other oil and natural gas deposits, brought Alberta prosperity, which lasted until the world drop in oil prices brought on the recession of the early 1980s. (R.B.D.)

Manitoba

Lying midway between the Atlantic and Pacific oceans, Manitoba became the fifth province of Canada when it was admitted to the confederation in 1870. Manitoba straddles the boundary between the prairie and central Canada. It has a large agricultural sector and a topography similiar to those of the provinces of Saskatchewan and Alberta; it also has a mixed economy, an urban orientation, and a multiethnic character, all of which resemble Ontario's. While other areas of the Canadian west have experienced economic cycles of boom and bust, Manitoba has maintained a steadier pace, missing sudden growth spurts and escaping abrupt collapses. Similarly, the province's political and cultural life has largely avoided the extremes that tend to characterize western Canadian society.

A relatively stable society

The province's name comes from an Indian word meaning "the god who speaks." More than two-fifths of its land area is forested. Its 250,947 square miles (649,950 square kilometres) contain 39,224 square miles of water, and, of its more than 38,500 lakes, Lake Winnipeg is the 13th largest in the world. The province is bounded on the north by the Northwest Territories, on the northeast by Hudson Bay, on the east by Ontario, on the south by the U.S. states of Minnesota and North Dakota, and on the west by Saskatchewan. Winnipeg, the seventh largest Canadian city, is the capital.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief and soils.* Extreme southern Manitoba is part of the Saskatchewan plain, a land of rich, level prairies and rolling pastures. The Manitoba Lowland to the north is the basin that once held glacial Lake Agassiz, remnants of which include Lake Winnipeg (9,416 square miles), Lake Winnipegosis (2,075 square miles), and Lake Manitoba (1,785 square miles). Upland plateaus, wooded river valleys, limestone outcrops, forests, and swamps mark the area that is drained by the Red and Assiniboine rivers into Lake Winnipeg.

To the north and east of the lowland is the geologically ancient Canadian Shield, an area of rocks, forests, and rivers. It covers three-fifths of the province and is drained by the Nelson and Churchill rivers into Hudson Bay. The Hudson Bay Lowland extends 100 miles (160 kilometres) inland as a flat plain of tundra and boglike muskeg. Soil conditions change from north to south. In the fertile zone black soils suitable for farming dominate, although there are occasional large sandy areas. North of this region, lower-quality black soil and gray-wooded soils are found. Soil in the shield, generally not suitable for agriculture, is characterized as gray wooded, podzol, and peat. Manitoba's Western Upland is on the Saskatchewan border. The Riding, Duck, and Porcupine mountains form the Manitoba Escarpment, whose highest point is Baldy Mountain at 2,727 feet (831 metres).

Temperature extremes

Climate. Manitoba has a moderately dry climate with sharp seasonal temperature changes. Temperatures of −40° F (−40° C) are not uncommon in any part of the province; summer days of 100° F (38° C) are not unusual in the southern regions. Average daily temperatures at Winnipeg range from 8.8° F (−12.9° C) in January to 79.7° F (26.5° C) in July. Average lows vary from −8.7° F (−22.6° C) in January to 56.8° F (13.8° C) in July. Annual precipitation varies from 14 inches (355 millimetres) in the north to 22 inches (560 millimetres) in the southeast, with about two-thirds of it falling between May and September. Snow typically covers the ground from November to April in the south, longer in the north.

Plant and animal life. The southern part of the province was once covered by open grasslands, most of which has been converted into farmland. The southern plains are relatively treeless, except for the river valleys, where stands of aspen, oak, willow, and poplar are common. Manitoba's more than 148,000 square miles of forest include open parklands of ash, Manitoba maple, elm, and oak in the south, becoming a mixed broad-leaved and coniferous forest to the north and west, and true northern coniferous forest in the higher latitudes. Trees in the north include white and black spruce, jack pine, aspen, white birch, and tamarack. The Hudson Bay lowlands have willow and spruce, as well as moss, lichens, and sedges.

Caribou, Arctic fox, marten, wolf, otter, lynx, red squirrel, and mink live in the northern forests; deer, moose, cougar, elk, black bear, beaver, weasel, raccoon, red fox, coyote, and muskrat inhabit the more southerly areas. Polar bears roam along Hudson Bay. Grouse, prairie chicken, and other game birds live in the uplands, and millions of geese and ducks breed in Manitoba's sloughs and ponds. Fish include bass, pickerel, sauger, pike, trout, and whitefish. Beluga whales inhabit Hudson Bay.

Settlement patterns. Manitoba's countryside has an older, more settled look than the other Prairie Provinces. The development of the Red River colony in the 19th century gave the early province a distinctly rural character. The rapid settlement of the agricultural lands between 1870 and 1914 and subsequent expansion into the Interlake and northwest areas maintained the rural dominance. During World War II, however, the rural–urban balance tipped. Since that time, the depopulation of rural Manitoba has continued apace, and Winnipeg and the province's network of smaller urban centres have grown accordingly.

Secondary towns

Other than Winnipeg, the province's main towns are Brandon, an industrial and agricultural centre serving the southwest; Thompson, a nickel-mining and nickel-processing town in the northern forest; The Pas, a trading and communications centre on the Saskatchewan River; Flin Flon, a mining centre near the Saskatchewan border; Churchill, a transshipment centre and port on Hudson Bay; Dauphin, a regional service town in west central Manitoba; and Selkirk, the centre of commercial fishing and water transportation on Lake Winnipeg.

The people. *Population composition.* Manitoba is home to a number of native peoples, including American Indian groups such as the Assiniboin and Ojibwa in the south, the Cree and Chipewyan in the north, and Inuit on the Hudson Bay coast. The native peoples have occupied the region for thousands of years, although historically there have been considerable population shifts. In addition, their numbers were greatly reduced as a result of exposure to European diseases, particularly smallpox. The métis (people of mixed ancestry, chiefly French and American Indian) developed a unique plains culture in the early 19th century. The early agricultural settlement in the Red River colony attracted a number of Scottish and Irish farmers. Nonnative settlement grew rapidly between 1870 and 1900, as a result of improvements to river and rail transportation. Immigration from Central Canada and, after 1900, from Europe continued until World War I. Manitoba's population declined during the depression in the 1930s and World War II but grew steadily thereafter, largely through natural increase.

Manitoba is noted for its ethnic diversity. There are numerous ethnic enclaves within the city of Winnipeg. The north end of town retains a strong Ukrainian character, and St. Boniface has one of the largest French Canadian communities outside Quebec. French Canadians established a number of communities south of Winnipeg, Germans settled in south-central Manitoba, and a sizable Icelandic settlement developed around Gimli, on the shores of Lake Winnipeg. The current population is more than 40 percent British, with significant ethnic minorities that include Germans, Ukrainians, French, native peoples, and Poles. The Roman Catholic and United churches are the largest religious denominations, although there are substantial numbers of Anglicans, Ukrainian Catholics, Greek Orthodox, Lutherans, and Mennonites, plus smaller groups.

In- and out-migration

Demographic trends. Mainly because of its faltering late 20th-century economy, there has been a steady out-migration of young people and professionals to other provinces. Manitoba's population growth rate stagnated during the late 1970s and early 1980s; the province enjoyed a slight net in-migration in the mid-1980s, but growth slowed again in the last part of the decade.

The economy. *Resources.* Manitoba has a range of resources to draw upon, allowing it to diversify its economic base. The province has more than 12 million acres (5 million hectares) of land in agricultural use, a huge network of streams, rivers, and lakes with considerable potential for irrigation and hydroelectric development, massive stands of timber, significant fish and game resources, and numerous mineral deposits, including nickel, copper, gold, lead, silver, cadmium, and zinc. Oil and natural gas have been found in the southwest, particularly around Virden, although the fields are relatively small.

Agriculture, forestry, and fishing. Agriculture remains an important sector of the Manitoban economy. Climate and soil types limit agricultural production; some marginal areas, including the Interlake District, have been farmed for several generations with little success. While much of the agricultural production is grain for export, a sizable market gardening sector has developed in the lands bordering the Red and Assiniboine rivers. In addition, poultry, hogs, and cattle are raised through the southern districts of the province. Wheat remains the province's main crop; others include barley, rapeseed (canola), flaxseed, and rye. Manitoba produces a variety of specialized crops, such as sugar beets, sunflowers, corn (maize), canning vegetables, and potatoes.

Foresty dominates northern Manitoba's economy. More than one-fourth of the province's landmass supports valu-

E. Otto/Miller Comstock Inc.

Field of sunflowers on a farm in the Assiniboine River valley near Carberry, Man.

able timber, although there has been some devastation from forest fires. Manitoba also has a specialized commercial fishing industry. Much of the fishing is done on the major lakes—Winnipeg, Manitoba, and Winnipegosis.

Industry. Manufacturing has overtaken agriculture as the largest sector of the province's economy. The traditional industries—meat packing, flour milling, lumber, pulp and paper, printing, and clothing—have been supplemented by more advanced processing and manufacturing. Food processing and distilling remain important, as do nickel and copper ore smelting and the manufacture of aeronautical systems, computers, electrical equipment, and a variety of other products.

Major industrial plants

Among Manitoba's major industrial establishments are a nuclear-research plant in Pinawa, an integrated forest-products complex in The Pas, a nickel-processing complex in Thompson, a pulp mill in Pine Falls, and a fertilizer plant in Brandon. Most of Manitoba's non-resource-based industrial capacity is located in Winnipeg. Manitoba Hydro, the government-owned utility, has been developing a number of large power stations along the Nelson and Saskatchewan rivers. Hydroelectric power is supplemented by coal-fired thermal plants at Brandon and Selkirk.

Winnipeg, once the commercial centre for western Canada, remains a major financial centre, home to the Winnipeg Commodity Exchange and other financial institutions. The growth of government services has also contributed to the expansion of the Manitoba service sector. Tourist travel grew steadily in the 1980s.

Transportation. Because of its location, Manitoba has long been a focus for transportation activity. The construction of transcontinental railways after Manitoba joined the confederation stimulated much development in the region. In outlying areas, branch lines are being abandoned as roads are built and passenger traffic declines. The province has an extensive network of major highways running east to west (including the Trans-Canada Highway) and north to south, supplemented by secondary gravel roads.

Winnipeg is a major hub for air travel; scheduled flights carry travelers to major Canadian centres and the United States, while nonscheduled operators provide service throughout isolated areas of Manitoba.

Churchill, a railway port and former military base on the shore of Hudson Bay, provides a short, direct sea route to Europe. The port is open for only 10 weeks in late summer, however. Waterborne freight traffic plys Lake Winnipeg in summer, and isolated northern communities and logging and mining camps are served by sled trains and vehicles designed for winter roads.

Administration and social conditions. *Government.* Manitoba derives its authority from the Canadian Constitution and the Manitoba Act of 1870. The lieutenant governor of Manitoba, appointed by the (federal) governor-general in council, represents the crown; the duties of this office are largely honorary and ceremonial. The functioning head of the government is the premier. The unicameral legislature sits for five years or until dissolved by executive action.

Executive branch

Municipalities are incorporated by the province, and local administration is under provincial control. Each municipality is governed by a council headed by a mayor or reeve. Districts are governed by provincial administrators, but there are no counties. Greater Winnipeg, consolidated into a single administrative unit in 1972, has an elected mayor and council. Provincial policing is handled by the Royal Canadian Mounted Police.

Education. Elementary, secondary, technical, and vocational schools and universities are a provincial responsibility. Schools are administered by elected local or larger-unit boards under supervision of the Department of Education and are financed by government grants and local taxes. After French-speaking Manitobans fought for their linguistic rights, the availability of French as a language of instruction was reinstated.

Manitoba has three universities. The University of Manitoba in Winnipeg is the major academic institution and home to most professional schools. The University of Winnipeg, noted for its Institute of Urban Studies, is a major cultural centre in downtown Winnipeg. Brandon University is small but noted for its music school and native studies program. Vocational schools are based in Winnipeg, The Pas, and Brandon.

Health and welfare. The people of Manitoba enjoy a relatively high standard of living and access to a wide range of government services. Sanitation and water-delivery systems protect against outbreaks of disease and are part of an extensive public health and safety service. The province supervises medical and hospital insurance programs and provides all Manitobans with free medical insurance. Manitoba has an extensive hospital network, with most specialized services available in Winnipeg or in regional centres. The province also maintains a substantial social welfare infrastructure.

Cultural life. Manitoba's diverse population has created a rich and varied cultural life. Ethnic groups—European, Asian, and native peoples—maintain vigorous attachments to their traditional cultures through schooling, dance, art, music, and crafts. The province's cultural diversity is not uniform. Winnipeg exhibits a variety of cultures, while the towns and villages of the southwest have a strong, uniformly British tradition. Ethnic enclaves throughout the province reflect the pattern of initial settlement. Manitoba is noted for its efforts to retain minority languages.

Cultural institutions

Manitobans have a reputation for supporting cultural agencies. Winnipeg is home to major artistic institutions: the internationally famous Royal Winnipeg Ballet, the Winnipeg Symphony Orchestra, Manitoba Theatre Centre, Rainbow Stage, Le Cercle Molière, and the Winnipeg Art Gallery. A major centre in downtown Winnipeg houses a concert hall and planetarium.

The province maintains a strong sense of its past. The Manitoba Historical Society is the leading agency in this regard. The Archives of Manitoba is a major repository for historical documents, including the collection of the Hudson's Bay Company. Museums and historic sites in Manitoba include the Manitoba Museum of Man and Nature in Winnipeg and the reconstruction of Lower Fort Garry, a National Historic Site north of Winnipeg.

Manitobans sponsor and support an exceptional array of community and cultural festivals. Among the more important are the Winnipeg Folk Festival; Folklorama, also in Winnipeg; Gimli Icelandic Festival; National Ukrainian Festival at Dauphin; Festival du Voyageur in St. Boniface; and Northern Manitoba Trappers' Festival at The Pas. Professional teams include the Winnipeg Blue Bombers in the Canadian Football League and the Winnipeg Jets in the National Hockey League. Curling is the premier winter sport. Other sports include hockey, baseball, cross-country skiing, fishing, and hunting. Manitoba has one national park, Riding Mountain, and numerous provincial parks.

HISTORY

The fur-trade era. Native peoples, who had lived in the Manitoba region for thousands of years, first came in contact with Europeans through the fur trade. Explorers searching for the Northwest Passage reached Hudson Bay in 1610, when Henry Hudson navigated the east side of the Bay. He was followed by a number of adventurers, including Thomas Button (1612), Jens Munk (1619), and Luke Fox and Thomas James (1631). At the urging of French Canadian adventurers a trading company, Radisson and Des Grosseilliers, was formed, which in 1670 received a monopoly over the fur trade in an area designated as Rupert's Land. The Hudson's Bay Company established a number of posts along the bay, facing tremendous competition from French traders. It was the French, led by Pierre Gaultier, Lord de La Vérendrye, and his sons, who first extended the trade into the interior; they reached the Red River in the early 1730s and established a series of posts in the area, including Fort Rouge on the present site of Winnipeg. The French trade cut severely into Hudson's Bay Company returns at its post by the bay and forced the company to expand inland. Several decades of intense competition ensued. The Montreal-based North West Company worked closely with the métis and resisted the British company's attempts to establish an agricultural colony at Red River. The bitter, often violent competition culminated in the Seven Oaks Massacre of 1816. Increas-

The Hudson's Bay Company

ing violence and declining profits forced the two firms to merge into the Hudson's Bay Company in 1821.

Settlement boom. The successful settlement of the western United States in the 19th century encouraged expansionists in Central Canada to look to the western regions of British North America. The agricultural potential of the vast prairie lands west of Red River was documented by the Palliser Expedition, commissioned in 1857 by the British government. Negotiations between the Hudson's Bay Company, the British government, and the government of the new Dominion of Canada resulted in an agreement in 1869 to transfer Rupert's Land to Canada. In 1870, under the Manitoba Act, the territory joined the confederation as a province; the remaining lands were designated the Northwest Territories.

The federal government encouraged western settlement, providing land to would-be settlers. There was an initial rush to the new province, but the absence of a rail connection to markets reduced the attractiveness of the region. Construction began on the Canadian Pacific Railway, which, by the early 1880s, had reached Manitoba. Settlement was delayed by a métis uprising in 1885 and was further slowed by the continued availability of homestead lands in the United States. Some settlers did come in this period, Ontario farmers, Mennonite immigrants, and Icelandic peoples among them. At the turn of the century, immigration to Manitoba boomed, fueled by massive government advertising, social conditions in Europe, and the decline in available land in the United States.

Era of prosperity

Manitoba in the 20th century. The immigration boom ushered in an era of prosperity and growth. Winnipeg grew rapidly and became the major urban centre for all of western Canada. Manitoba farmers, aided by reduced freight rates, higher world prices for wheat, and improved strains of grain seed, enjoyed unprecedented prosperity. The economic boom ended just before World War I, ushering in a depression that lasted through the first years of the conflict. Labour unrest over wages and working conditions arose during and peaked after the war.

Manitoba's economy was transformed during the early part of the century. A strong agricultural sector, diversified among wheat and other grains, cattle, and market gardening, provided the province with a measure of stability. Winnipeg's industrial economy, however, grew rapidly. Manitoba also achieved its current form, expanding westward in 1881, eastward in 1884, and northward in 1912. The province was severely affected by the depression of the 1930s, although the problem was more one of prices than of production in the agricultural sector. Nonpartisan or coalition governments ruled the province from 1922 to 1958, when Dufferin Roblin led the Conservatives into office. The return to prosperity during World War II was matched by a resolve to diversify the province's economic base. The provincial government gained control of natural resources in the 1930s and began to encourage northern development, particularly of mining and lumbering. Government involvement in the economy peaked during the administrations of the New Democratic party from 1969 to 1977 and 1981 to 1986, with major investments in infrastructure and some controversial experiments in direct ownership. (P.McL./K.S.C.)

Saskatchewan

Saskatchewan is one of only two Canadian provinces without a saltwater coast, and it is the only one whose boundaries are wholly artificial. It lies between the 49th and 60th parallels of latitude and is bounded on the west by longitude 110° west of Greenwich, and its eastern limit, with minor adjustments, is longitude 102°. Its southern half is largely an extension of the Great Plains of central North America, rarely rising 2,000 feet above sea level, and its northern half, most of which lies in the ancient rock mass of the Canadian Shield, is sparsely populated bush country with many lakes and tundra. Its area is 251,866 square miles (652,330 square kilometres), of which 31,518 square miles are water; and it measures 760 miles (1,223 kilometres) from south to north, tapering from a width of 393 miles (where it abuts Montana and North Dakota in the United States) to 277 (where it meets the Northwest Territories). In area, Saskatchewan is Canada's fifth largest province, and in population, its sixth. Economically, the province has always been heavily dependent on the exportation of its agricultural and mineral products and is thus peculiarly sensitive to fluctuations in world markets beyond its own or even Canada's control.

Dependency on agriculture and mining

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* Although familiarly known as one of the Prairie Provinces, Saskatchewan has little native prairie; a large proportion of its productive acreage (half the province) is rolling ranch and parkland, both of which offer immense vistas from their higher points. There is not a single mountain in the province, although the term is loosely used to identify several landmarks. The Cypress Hills, in the southwestern corner of Saskatchewan, include the provincial summit: 4,816 feet (1,468 metres) above sea level. The hills constitute the only part of the area to escape glaciation and contain unique plant and animal life. Wood Mountain (3,275 feet) and the Vermilion Hills (2,500 feet) are some of the province's other major departures from the rolling plains topography. Cut into the plains are many spectacular river valleys, the most notable being those of North and South Saskatchewan and the Qu'Appelle.

Drainage and soils. Saskatchewan drains from west to east, its great rivers (which provided the first transportation routes) rising in the Rocky Mountains and emptying ultimately into Hudson Bay. The soils through which the rivers flow are predominantly chernozemic (dark-coloured grassland) and podzolic (light-coloured forest) with extensive deposits of poorly drained mineral and peat soils in the north.

Climate. The climate keeps much land out of agricultural production. In the southern half of the province there are as few as 80 to 100 frost-free days annually. Temperature variations are extreme; January temperatures have fallen below −65° F (−53° C) in settled parts, and in July temperatures upward of 105° F (41° C) have often been recorded. The normal mean daily reading for the arable regions ranges from −5° F (−21° C) to 10° F (−12° C) in January and from 55° F (13° C) to 65° F (18° C) in July. Precipitation generally is not high, averaging from 10 to 20 inches (255 to 510 millimetres) each year; snowfall ranges from 30 inches (760 millimetres) in the southwest to more than 60 inches (1,525 millimetres) in the north-central area. Drought years are not uncommon.

Botanical regions

Plant and animal life. Saskatchewan from north to south is marked by six recognizable bands of natural plant life, all running in a southeasterly direction and roughly following the pattern of soil deposits. Farthest north is the subarctic forest tundra, south of which lie the northern coniferous forests and then a strip of mixed forest. The most northerly agricultural belt is aspen parkland, parts of which are still being cleared. The two most southerly bands are composed of midgrass prairie and shortgrass prairie. The three most southerly zones produce a rich profusion of attractive wildflowers, many of which also, paradoxically, qualify as noxious weeds.

The vast unsettled parts of Saskatchewan support a large wildlife population of great variety. Grizzly bears and mountain lions are now rare, but wolves, black bears, moose, deer, caribou, elk, and antelope are common, together with enormous numbers of smaller mammals. Coyotes, foxes, and lynx, together with the gophers, rabbits, and other creatures they prey on, are abundant, and the province supports a considerable amount of trapping. Saskatchewan is on the main western flyway of waterfowl, songbirds, hawks, and owls, many of which nest in the province. The extensive water resources maintain both commercial and game fish in quantity. Northern Saskatchewan, particularly, is a haven for the hunter and angler.

Pollution. The lack of heavy industry and of metropolitan areas keeps Saskatchewan relatively free of the kinds of pollution associated with high population density and manufacturing, but the extensive agricultural development subjects it to those connected with weed killers and insecticides. Significant amounts of mercury have been found in fish and birds, and continuing research suggests that the amount of contamination in wildlife may be larger than had been apparent. The sources of the major rivers also subject Saskatchewan to upstream pollutants from areas over which it has no control; but prevailing winds do not come from heavily polluted regions, and the air is generally clean, though an occasional northern forest fire casts a pall over thousands of square miles to the south. Frequent strong winds produce dust clouds.

Settlement patterns. Saskatchewan's landscape makes its inhabitants conscious of the sky; and the changing patterns of light and shadow on clouds, which commonly offer magnificent sunrises and sunsets, are as much a part of the scenery as any contour of the earth. All of Saskatchewan is farther north than any of the most densely populated parts of Canada, and the province's own north is largely inaccessible except by air, with the result that few citizens are familiar with it. Saskatchewan's best-known regions and sites are its main agricultural and recreational areas: the wheat belt, the ranching country, the Qu'Appelle valley, the Cypress Hills, Lake Diefenbaker, Waskesiu Lake, the old fur-trading routes and trails and their inevitable forts, and the sites of Saskatchewan's few battles.

Saskatchewan is not dominated by any metropolitan centres. Unlike most Canadian provinces, Saskatchewan has in effect two capitals: Regina, the official capital, in the central south, and Saskatoon, 160 miles to the north. Both cities are growing rapidly. Other major cities include Moose Jaw and Prince Albert. Because the four largest cities maintain less than 50 percent of the population, a small-town character is maintained. The most striking man-made feature of the landscape has been for decades the grain elevator, and the typical village is clustered around three or four of these structures. But changing technology has been making the elevators obsolete, and they are being replaced by facilities set farther apart.

Major population regions

The people. The population has changed markedly during the area's history. It was originally exclusively American Indian, to which French and British elements were added early, and a large population of mixed origin, the mêtis, developed. Then other European groups—German, Austrian, Ukrainian, Scandinavian, Russian, and Polish, among others—were attracted, some by generous homestead grants, and some in part by a desire to avoid compulsory military service in their own countries. The period of heaviest immigration was between 1900 and 1920, when the population rose suddenly from less than 100,000 in 1901 to nearly 700,000. The population of British origin was, by the late 20th century, less than 40 percent. Many of these groups, including the British, settled in separate communities where they could use their own language and continue their own religion and customs, and Saskatchewan contains many settlements readily identifiable as Ukrainian, French Canadian, or German. Since the 1960s an appreciable number of migrants from India and the Far East have arrived, settling mainly in the cities. Provincial law permits the use of languages other than English in schools for specific purposes, and widespread advantage has been taken of the law. Ethnic variety is matched by that found in religious affiliation; the largest churches are the United Church of Canada, Roman Catholic, Lutheran, Anglican, Ukrainian Catholic, Greek Orthodox, Mennonite, Presbyterian, and Baptist.

Population growth has been generally slow because of the high emigration rate, which exceeds immigration. This is attributed to increases in the efficiency of Saskatchewan's basic industries, which can steadily produce more with less manpower. Internally, the province, although still one of the least urbanized, has one of the highest rates of urbanization in Canada, the growth coming chiefly from the movement of rural dwellers into urban areas.

The economy. Saskatchewan's economy, since its beginning, has been based on extractive industries: forest products, fish, furs, agricultural products, oil and gas, and potash. In almost all cases the products are consumed outside the province and generally outside Canada, a situation that makes the province one of the most economically vulnerable areas in the world. A grain belt, made up predominantly of wheat but also including large acreages of barley, oats, rapeseed (canola), flax, and rye, lies between the southern border and the 54th parallel of latitude. Potash is found in a narrower band running diagonally across the province from west to east, its northernmost point being west of Saskatoon. Oil and gas lie in the southernmost quarter of the province, while there are rich uranium deposits in the north; there also diamonds of industrial quality were discovered in the 1980s. Other significant minerals include salt, sodium sulfite, lignite, zinc, copper, gold, and a variety of clays. In terms of employment, manufacturing has always played a relatively minor role in the provincial economy and is characterized by several hundred small establishments, most of which have only a few dozen employees.

Minerals

The Saskatchewan economy, always dependent on external markets, has internally required a variety of governmental supports; as a result the province has never had a true free-enterprise system, while public enterprise and mixed public and private ventures have characterized the development of the economy from the beginning. The Canadian protective tariff, long criticized by prairie dwellers because it made them pay more for goods of all kinds, appears to be the mainstay of the developing secondary industries. The first waves of settlers, attracted to Saskatchewan by federal policies, were carried on railways built with federal assistance. Saskatchewan grain moves to federal terminal elevators at controlled freight rates. Within the province, political parties in power, regardless of ideology, have sponsored and maintained publicly owned utilities, a bus company, an insurance company, and public hospitalization and medical care. Nonetheless, in the 1960s and again in the 1980s the people elected governments that stressed private capitalism.

Transportation. Modern Saskatchewan was originally the creation of transcontinental railroads, which carried settlers and supplies in and grain out. Though freight remains an important rail component, passenger carriage has declined, and services have been reduced or abandoned. The province is now crisscrossed with highways. The system of land division in the rural areas provides for "road allowances," strips of territory a mile or two apart that serve as simple, mostly dirt, roads, which when dry are firm and passable and widely used for local travel.

Except for recreation, water transportation is all but obsolete in Saskatchewan; small shallow-draft steamers formerly sailed the main rivers, but, since the rivers are shallow with shifting sandbars, they have not been sig-

George Hunter

Harvesting wheat on a Saskatchewan farm in the grain belt near Saskatoon. A potash mine appears in the distant background.

nificant transportation routes since before World War I. Airlines, by contrast, have developed dramatically in Saskatchewan, where approximately half the province is accessible only by air. The small plane serves the north for both commerical and recreational purposes, and all major centres are on scheduled airlines.

The provincial constitution

Administration and social conditions. Saskatchewan's constitution, based on custom and the Saskatchewan Act of 1905, provides for a British parliamentary system, in which the life of the executive depends on the support of a majority in the legislature. A general election must be held every five years; short of that period, the premier may advise the holding of an election at any time, and most assemblies last only four years. As in all the provinces, the lieutenant governor is appointed and has become by custom and judicial decision the counterpart of a constitutional monarch, whose position and powers are largely symbolic. Saskatchewan's larger centres have their own local police, but in the province as a whole the law is enforced by the Royal Canadian Mounted Police.

Unlike those of Canada's other Prairie Provinces, the Saskatchewan legislature has a long tradition of strong vocal opposition in the assembly, with the Liberals and the New Democratic Party (NDP; formerly the Cooperative Commonwealth Federation) traditionally providing what was a two-party system ideologically divided into free enterprise versus democratic socialism. Since the 1970s the Progressive Conservatives have gained support at the expense of the Liberals, and they became the governing party in 1982, ousting the NDP. Like the Liberals, the Progressive Conservatives draw their greatest strength from rural areas; the New Democrats have a stronger urban base. Whichever party is in power is vigorously opposed by the others, with the Liberals and Progressive Conservatives generally espousing development of the province by business and corporate means and the NDP strongly urging the use of public and cooperative enterprise.

The province is divided into a multiplicity of local administrations including hospital districts and school districts, all constitutionally under provincial jurisdiction but all having considerable local responsibility. Municipal government in Saskatchewan is based on the U.S. mayor–council model, with a mayor elected separately from the council, and a number of appointed boards and commissions operating largely independently of either.

Federal social aid

For most of its history the province has qualified for the kinds of federal aid available to those whose economy operates below the national average. The province's reliance on federal subsidies as a percentage of total revenues, though it varies with crop conditions, is generally above the national average. Saskatchewan's wage levels for both industry and agriculture are never among the lowest for the provinces but are neither among the highest. The province's "middle" position carries over into its internal affairs: it is socially and economically (except for its poor American Indian and métis peoples) one of the least stratified areas in Canada, having little of great individual or corporate wealth on the one hand, and little general destitution on the other.

Cultural life. Although lacking great metropolitan centres, Saskatchewan has developed creditable art galleries and professional theatre and opera companies; even so, many artists leave for careers elsewhere. Writing in and about the province, always strong, has blossomed since the 1960s, and the same is true of painting and sculpture. Saskatoon and Regina have excellent civic auditoriums, and Saskatoon has an outstanding branch of the Western Development Museum, whose chief exhibits, outmoded farm machinery and automobiles, are annually refurbished in a celebration of pioneer days. The province is well served by the radio and television networks of the Canadian Broadcasting Corporation, augmented by private broadcasting services.

The province's oldest recognizable cultural institution is the University of Saskatchewan, established in 1907 and, with remarkable foresight, given a huge campus at Saskatoon. The university has produced much fundamental research relevant to Saskatchewan (it maintains, for example, an Institute of Space and Atmospheric Studies that has done extensive research on the aurora borealis) and has also a tradition of extension services on and off campus. It has sent forth a steady stream of distinguished workers in a variety of fields, from theatre to nuclear physics. As a result of the limited opportunities available, not many of these have been able to remain in the province. The province is also noted for the number of professional hockey players and curling champions it has produced.

HISTORY

Early exploration

Since Saskatchewan became a full member of the Canadian federation only in 1905, much of the area's historical interest depends on events vastly older than the province. Dinosaur and mammoth finds have been common. The first known human inhabitants were American Indians of several linguistic groups who were present at least 5,000 years ago; they were mainly hunters. With the coming of the Europeans, they became trappers involved in the fur trade. The first European known to see the Saskatchewan River was Henry Kelsey, who in 1691 explored part of the plains for the Hudson's Bay Company, which received its charter in 1670 and is still extant. Fur traders and buffalo hunters, variously American Indian, métis, French, and British, and explorers and missionaries made up the bulk of the area's inhabitants until the second half of the 19th century.

Acquisition by the Dominion of Canada

The area from which Saskatchewan is carved was first granted to the Hudson's Bay Company and then, in 1869, surrendered by the Rupert's Land Act back to the British crown, in order that it could be turned over to the newly formed Dominion of Canada, which was done in 1870. Canada administered its newly acquired western territories almost as if they were colonies and, in 1873, created the North West Mounted Police to maintain law and order. In 1885 the national authorities sent out troops to quell the second Riel Rebellion, an uprising in which a large number of métis, by now deprived of their main sustenance, the buffalo, sought to establish their rights to western lands in the face of growing settlement. Constitutionally, the territories in 1875 were granted an executive council with a promise of an elected assembly, and by 1897 they had won responsible parliamentary government on the British model.

Saskatchewan, created by the Saskatchewan Act in 1905, entered confederation with its present boundaries and the status of a province equal to the others except that, as with its sister province Alberta, the federal government retained control of its natural resources, paying a subsidy in place of the revenues the resources might have yielded. (The resources were assigned to the province in 1930.) The new provincial government, after a good deal of rivalry among the towns, chose Regina, the former territorial capital, as its centre of operations, and the first premier appointed was Walter Scott, a believer in partisan politics, as opposed to those who favoured a continuation of the kind of cooperative effort that had led to the creation of Saskatchewan as a separate province. A member of the party in federal power at the time, the Liberal Scott was the first of several able politicians who kept the party in power in Saskatchewan except in 1929–34 and 1944–64, and after 1971. The 1944–64 period was unique in North American history. During that era the Cooperative Commonwealth Federation (CCF), successively led by T.C. Douglas and Woodrow Lloyd, established the first avowedly socialist government on the continent, and the party won international attention in 1962 when it implemented the continent's first compulsory medical care program, accompanied by a doctors' strike.

Privatization and co-ops

Regardless of which political party has been in power at any given time, the Saskatchewan environment has always demanded much governmental intervention in the economy. The provincial telephone company and the power and gas utility, for example, were publicly owned (although neither was created by a socialist government) down to the 1980s, when privatization began under a Progressive Conservative government. The cooperative movement has been encouraged by all parties and has been influential in a wide range of service, retail, and wholesale activities that include large credit unions and an oil refinery. In the handling of grains, the backbone of the province's economy, the Saskatchewan Wheat Pool, also a cooperative, has been a dominant influence. The co-ops helped many individuals survive the drought and depression of the 1930s, during which Saskatchewan society is considered to have sustained setbacks as severe as any suffered in Canada. After World War II the province attained a major development in mineral exploitation and industrial growth, and its diversified base was combined with new farming techniques to strengthen the economy. (N.W.)

BRITISH COLUMBIA

One of the last regions of the North American continent to be explored and settled, British Columbia emerged in the second half of the 20th century as one of the leading provinces of Canada in population, economic wealth, and overall growth potential. Its 365,947 square miles (947,800 square kilometres) of total area make it the second largest province in physical size. The land has a diversity of climate and scenery unparalleled in the nation, from the island-studded and fjord-indented coast to the great peaks of the western continental cordilleras with their large interior plateaus. The province is bounded on the west by the Pacific Ocean and the panhandle region of Alaska, on the north by the Yukon and Northwest territories, on the east by Alberta, and on the south by the U.S. states of Montana, Idaho, and Washington.

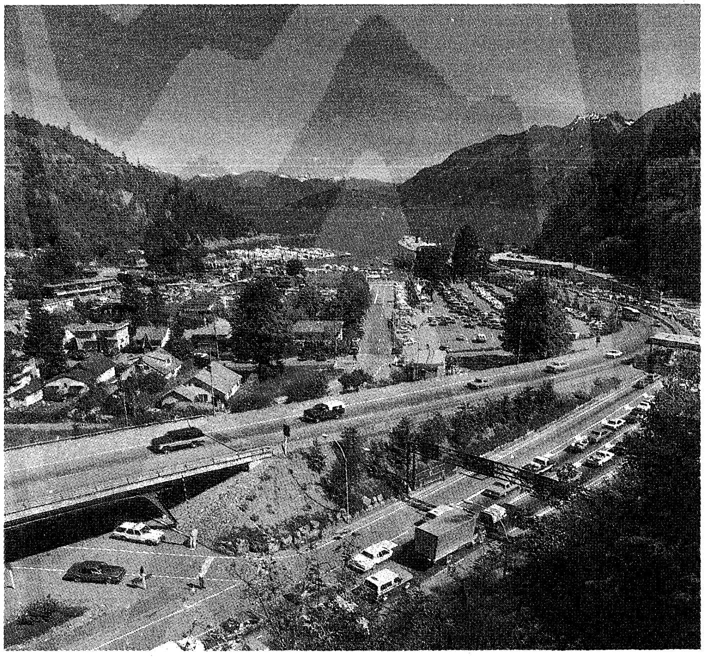
British Columbia's relative prosperity and high growth rate have made it the third most populous province, with a population exceeding three million. Its main cities include Vancouver, the largest port of Canada and of western North America, and Victoria, the provincial capital, located on the southeastern tip of Vancouver Island. British Columbia is one of Canada's more prosperous provinces.

PHYSICAL AND HUMAN GEOGRAPHY

The mountains, the sea, and the islands

The land. *Relief.* The vast territory of British Columbia lies almost entirely within the great mountain system, or cordillèra, that stretches along the western edge of the Americas from above the Arctic Circle to Cape Horn, at the southernmost extremity of South America. These mountains divide the province in ranges aligned in a southeast–northwest direction, creating a series of valleys and a broad central interior plateau where human settlement has concentrated. Hundreds of coast-hugging islands—the largest of which are Vancouver Island and the Queen Charlotte Islands—offer a protected waterway along the coastline, which is indented by narrow fjords that twist inland about the bases of towering mountains. The broad Fraser delta, behind Vancouver to the south, is the largest of the limited coastal lowlands, while in the interior many of the wide plateaus are cut by deep canyons and entirely surrounded by mountain ranges.

Brian Milne/First Light



The Coast Mountains rising over Howe Sound at Horseshoe Bay, West Vancouver, B.C., where (foreground) the Trans-Canada Highway connects with ferries to Vancouver Island and the final stretch of the highway.

Drainage. The province contains three main river systems: the Peace in the far north; the Fraser, which drains nearly all of the interior plateau; and the Columbia in the southeastern and south central regions. Lesser rivers, such as the Skeena, Nass, Iskut, and Stikine, drain the northwestern region into the Pacific, while the Liard system drains the northeastern section into the Arctic Ocean.

The Fraser, the only major river that lies entirely within the province, rises in the Rockies near the Yellowhead Pass, flows north and then southwestward to Prince George, where it turns almost due south for 300 miles (480 kilometres), flowing to Hope, and then westward through the lush farmlands to the sea south of Vancouver. The Columbia follows the Rocky Mountain Trench northward, bends around the northern end of the Selkirk Mountains, and turns south to flow into the Arrow Lakes and then into the states of Washington and Oregon. The Peace also becomes a formidable stream within the Rocky Mountain Trench, but it cuts eastward through the Rockies and into the plains area of Alberta.

Most of the thousands of lakes are small, but they are important for the water they store in an age when hydroelectric power has become a prized resource. The larger lakes are made long and narrow by the north–south mountain ranges that confine them in the deep and narrow valleys in all parts of the province. Atlin and Teslin in the northwest extend into the Yukon. Babine, Stuart, Shuswap, Quesnel, and François, which range from 90 to 200 square miles in area, are important salmon-spawning lakes. The Arrow and Kootenay lakes are important as storage reservoirs for hydroelectric plants.

Precipitation and temperature variation

Climate. Because of the Kuroshio, or Japan Current, which warms the coast, and the adjoining mountain ranges, British Columbia experiences a variety of climates. Some climatologists have claimed that, in its temperature, humidity, and variability, the southwestern corner of the province, tempered by the current, has one of the most favourable climates for people, plants, and animals. The prevailing winds from the Pacific, flowing over succeeding mountain ranges, cause a wide variety of precipitation and temperature ranges across the province, but along the coast such variation is negligible. Summers are comfortably cool, while winters are not severe, and 0° F (−18° C) is seldom recorded. In the Okanagan and Cariboo regions to the east, wider variations are recorded; summers are hot and winters are colder. Still farther east, up against the Rockies, similar temperature ranges prevail but with considerably heavier snowfalls. In the northern interior and Peace River country, very cold winters and very hot summers are normal.

British Columbia receives considerable precipitation, but its distribution is far from ideal. Some coastal towns have average annual rainfalls of 160 inches (4,100 millimetres) or more (perhaps the wettest region on the continent), while Ashcroft and Merritt, only 125 miles from the coast, register only 7 and 9 inches (180 and 230 millimetres), respectively. Victoria, sheltered by Vancouver Island mountains, receives less than 30 inches (760 millimetres), but Vancouver, 72 miles across the Strait of Georgia, receives 60 inches (1,525 millimetres), delivered by the winds crossing that body of water.

Plant life and soils. About 60 percent of the land area is forested, while only 3 percent has soil suitable for agriculture. The most valuable soils of British Columbia are the alluvial soils that developed on sand and silt laid down by streams and rivers. These extremely fertile soils, located mainly in the lower Fraser valley, are of limited extent. Distinctive areas of nonforested open grassland along the Fraser River, south of Williams Lake, and in the Kamloops–Meritt region south of the North Thompson River have rich pedocal soils upon which British Columbia's ranching industry flourishes. Similar prairie grassland soils lie in the Peace River country, the only part of the province suitable for large-scale grain farming. The more extensive but much shallower podzol soils cover the wet areas, especially along the coast, where they sustain a dense forest cover. The mild, wet climate of the coastal region produces large trees, some of the old growth specimens exceeding 200 feet (60 metres) in height. Western hemlock, red cedar, balsam fir, and Sitka spruce predominate, but the Douglas fir, which makes excellent sawn lumber, also grows in slightly drier coastal locations. The trees in Interior forests are smaller. They include Douglas fir and lodgepole and ponderosa pine in the south; balsam fir, spruce, and lodgepole pine in the centre; and white spruce in the far north. One-third of the province consists of barren Alpine tundra, snowfields, and glaciers; more than three-fourths is above 3,000 feet in altitude, beyond the limits of agriculture and often useful only for grazing.

Importance of the valleys

Settlement patterns. Because of the mountains and waters that surround them, British Columbians have become a society of valley dwellers. Valleys have assumed enormous importance for transportation, communications, and settlement, especially in the southern part of the province, where four-fifths of the population lives.

The mountains of Vancouver Island are separated from the Coast Mountains on the mainland by the Strait of Georgia, a saltwater valley along whose rocky slopes, on both sides, are located the province's two largest cities, Vancouver and Victoria. East of that valley is located the valley and canyon of the Fraser, with such service centres as Hope, Boston Bar, Lytton, and Lillooet, while still farther east lie the irrigated fruitlands of the Okanagan valley and the cities of Salmon Arm, Enderby, Vernon, Kelowna, and Penticton. North of the Okanagan, the landscape opens out into broad, gently rolling lands that form a basin surrounded by higher mountain ranges. Rapid expansion of the forest industry after World War II in the Central Interior plateau fostered growth of two regional centres, Kamloops and Prince George. Beyond and pressing against the Rockies are the parallel mining and lumbering valleys of the West and East Kootenays, which include the communities of Rossland, Trail, Nelson, Fernie, and Cranbrook.

For the most part, trails, roads, railroads, oil and gas pipelines, and electric power lines follow the routes of the rivers and often lie within yards of each other.

The people. *Population composition.* British Columbian society is one of the more British of Canada's 10 provinces, but it is also one of its most racially diverse. The English, Scottish, and Northern Irish played the major role in founding the province, and they still form the controlling elite. Scots dominated the fur-trade settlements, and when more than 25,000 miners arrived from California in 1858 the British government dispatched officials and the Royal Engineers to administer the colony. With the return of most American miners to California, British Columbia became essentially British. Immigrants from the British Isles flooded the province in the early years of the 20th century, reinforcing its ethnic character. By contrast, French Canadians are only a small minority scattered throughout the province. The arrival after 1945 of large numbers of Europeans, especially Dutch farmers attracted to Fraser delta lands, Germans to lower mainland cities, and Italians to various construction projects, challenged the province's Anglo-Saxon identity, but not nearly so forcefully as the presence of Asian minorities.

Chinese immigration

Chinese labourers first arrived at the time of the gold rush and suffered official discrimination from the 1870s, when they were disenfranchised. The large number of Chinese men imported in the 1880s to build the Canadian Pacific Railroad were joined later by additional Chinese, East Indian, and Japanese workers. As the Asian population increased to 9 percent of the provincial population in 1881 and 10 percent a decade later, anti-Asian feeling intensified. Discrimination peaked in 1923, when the Canadian government passed the Chinese Immigration Act, virtually ending Chinese immigration for a generation, and in 1942, during World War II, when all people of Japanese origin, whether Canadian citizens or not, were evacuated from the British Columbia coast. Due to more liberal attitudes about race, the declining population of Chinese men, and the forced dispersal of Japanese families, prejudice in British Columbia lessened after the war. Canadian immigration policies changed very slowly, however, and until 1967 continued to discriminate against racial minorities. Thereafter, the Asian population in British Columbia rose quickly through immigration from a low of 1 percent of the provincial total in the 1950s to more than 5 percent in the 1980s. Migration from Hong Kong during this period predominated, a movement fueled by fears of economic change that would accompany the scheduled takeover of the colony by China in 1997. A smaller but important movement of Sikhs from India to British Columbia in the mid-1970s added to the province's ethnic diversity, as did the arrival of Fijian and Vietnamese immigrants.

Among British Columbia's other minority groups, the most significant, for historical reasons, are the native Indians. It has been estimated that native people numbered about 80,000 in the late 1700s, before tuberculosis, diphtheria, smallpox, and other diseases brought by colonists reduced their ranks to 23,600 by 1934. Improved health measures inaugurated since then have permitted their numbers to increase to about the level attained two centuries earlier. Nonetheless, they now constitute less than 3 percent of the provincial population. Scores of reservations dot the province's landscape, chiefly in rural areas. The pressure of an increasing population combined with poverty began to drive Indians into the cities. Lacking education or skills valued by white society, they were relegated largely to underemployment and unemployment. But this relationship of dependence has been changing. Many native men and women now attend provincial colleges and universities and are fighting in the courts to reclaim traditional rights to land and resources taken from them without compensation.

U.S. immigrants

Although persons of U.S. origin constitute a small minority, they are influential in the economic life of the province—in manufacturing, mining, ranching, and forestry. Many have become part of the teaching, administration, and research staffs of the universities and colleges. About 2,500 blacks make up a small minority.

Religion. Religious practice has reflected the province's social character. European immigrants created a predominantly Christian society, and most native people converted to Christianity after the settlers arrived. In addition, the vast majority of British Columbians belong to one of many Protestant denominations, in contrast to the predominance of the Roman Catholic church that is found in other provinces. Religious conflict over whether schools should be denominational or nonsectarian never strongly influenced provincial politics in British Columbia, as it did elsewhere in Canada. The view held by mainstream Protestants that the church should remain separate from the state has been accepted as the norm in public affairs. In the early 1960s the United Church of Canada, formed in 1925 by Methodists, Congregationalists, and some Presbyterians, claimed the affiliation of almost one-third of British Columbians, followed by Anglicans and Roman Catholics. But British Columbia has also been the most secular of Canadian provinces, having the highest proportion of citizens who claim no religious affiliation. The province's Protestant identity is challenged as well by the ethnic diversification of its population. British Columbia is now home to almost one-fourth of Canada's adherents to the Eastern religions.

Demographic trends. British Columbian society has always been relatively fluid, and only recently has the proportion of the population born inside the province come to exceed that born elsewhere. Indeed, British Columbia has attracted immigrants more consistently since confederation than any other Canadian province, and in the

20th century more than any except Ontario. Interprovincial migration, especially from the neighbouring Prairie Provinces, increased noticeably in the 1930s. In the following decades British Columbia's pleasant climate drew retirement-age migrants to Victoria, Vancouver, and the Okanagan, while prosperity and high wages attracted young workers to construction and forest industry jobs in the central and northern part of the province. British Columbia is one of Canada's most urbanized provinces, with more than three-fourths of its residents living in urban places.

The economy. The economy of British Columbia is a mixture of public and private enterprise. Service industries have become the largest sector of the province, followed by manufacturing.

Resources. Natural resources have formed the basis of British Columbia's economy since the arrival of Europeans. Mining and mineral processing account for 3 percent of its labour force but nearly 20 percent of the production value of British Columbia's major industries. Mines are located throughout the province and include open-pit coal mines in the southeastern and northeastern corners of the province and open-pit copper mines southwest of Kamloops. Oil and gas have been drawn from wells in northeastern British Columbia since the 1950s, and new reserves continue to be found. Forestry and fishing are other major resources.

Mining

Agriculture, forestry, and fishing. Agriculture plays an important part in the province's economy, although it is not homogenous and varies widely, from the highly capitalized dairy industry of the lower Fraser valley to the fruit orchards and vineyards of the Okanagan, the mixed farms of the Bulkley valley, and the highly specialized grain farms of the Peace River country. Geography severely limits agricultural production in British Columbia, where cultivated land as a percentage of total provincial area is the second lowest in Canada.

Forestry is the most significant of British Columbia's resource industries. Traditionally centred on the west coast, where the dense rain forest has produced large trees easily accessible by water, logging, sawmilling, and pulp and paper production spread to the central and eastern parts of the province in the postwar years. Since 1972 the total volume of timber cut in the Interior has exceeded that on the coast. British Columbia has 17 percent of Canada's total forested land but 40 percent of its marketable timber, and by the late 1980s British Columbia was producing 70 percent of the country's sawn lumber and 90 percent of its soft plywood. The corporate structure of the industry also changed after World War II, with a few giant multinational corporations, often controlled by American capital, replacing the previously decentralized ownership. Corporate concentration of the forest industry also occurred as a result of provincial government policy.

The five species of Pacific salmon provide the foundation for British Columbia's sport and commercial fishing industry, although halibut, cod, herring, and various shellfish are also much sought after. Fish processing, which was carried on in dozens of small canneries spread along the coast, has since the 1940s become concentrated into a few large plants near Vancouver and Prince Rupert.

Electricity production

Power and manufacturing. The production of hydroelectric power has greatly facilitated British Columbia's economic expansion. Coal from Vancouver Island and, starting in 1898, from the Crowsnest Pass in the province's southeastern corner provided the major energy source for railroads and industry well into the 20th century. But British Columbia's mountainous relief and, in certain areas, high rainfall create vast potential for the production of hydroelectric energy. The quest for cheap electric power, which had been generated in relatively small quantities at several locations starting at the end of the 19th century, led to a privately funded project to produce electric energy for an aluminum smelter at Kitimat on the northwestern coast in the early 1950s and vast construction projects on the Columbia and Peace rivers in the 1960s. On the Peace the province's publicly owned electric utility, the British Columbia Hydro and Power Authority, erected the W.A.C. Bennett Dam, one of the world's largest earth-filled structures, and on the Columbia the Mica, Hugh Keenleyside, Revelstoke, and Duncan dams.

British Columbia's urban areas mainly provide services for, or process the products of, nearby salmon and other types of fishing, farming, mining, and logging industries. For instance, Nanaimo, once a grimy coal-mining town, now prospers as a market centre for fish, lumber, and pulp and paper. Trail specializes in mineral smelting, and Prince George and Kamloops are centres for the large Interior forest products industry. Kelowna, an agricultural service centre, also benefits from logging and mining as well as tourism. Resource industry management for the coastal region gravitated after 1900 to Vancouver, as did sawmilling, fish processing, and secondary manufacturing. Yet greater Vancouver, where 40 percent of the province's population resides, is less tied to the resource economy than other urban places in British Columbia. That area's multifaceted urban economy is increasingly bound by connections of trade and finance to markets around the Pacific Rim. The provincial capital and second largest city, Victoria, has become largely a service community.

Transportation. Although an elementary system of transportation was established in gold-rush days and most of its routes are still followed, no concerted effort was made to tie the isolated areas of the province together by roads, ferries, and railroads until after World War II. Highways are exceedingly important in a province fractured by towering mountains, long coastal inlets, and swift rivers. This crucial fact was recognized and exploited by the infant Social Credit Party, which, upon winning control of the government in 1952, adopted an ambitious and controversial program of road building, tunnel and bridge building, and ferry services as the chief plank of its highly successful political platform.

Vancouver Island, with one-sixth of the province's population, represents a special transportation problem. This has been met by the development of one of the world's largest ferry fleets. Most of the ferries connecting Vancouver with Victoria and Nanaimo carry 1,000 passengers and from 140 to 200 cars each.

Railroads

The province is served by four railroads. The Canadian Pacific Railway (CPR) connects Vancouver with eastern Canada and maintains a network of branches serving the mining, forest, and agricultural industries throughout southern British Columbia. The Canadian National Railway serves the south and the north with terminals at Vancouver and Prince Rupert. The Burlington Northern serves the province from the United States. The British Columbia Railway links Vancouver in the southwest with Prince George and Fort St. John in the north. The vastness of the province combined with difficult terrain has also encouraged the development of air travel, and several small companies now connect the far reaches of British Columbia with its urban centres. Vancouver International Airport is the main air transport centre.

Administration and social conditions. *Government.* Parliamentary government in British Columbia dates from the inauguration of the first legislature of the Colony of Vancouver Island, Aug. 12, 1856, but responsible government was not achieved until confederation.

The lieutenant governor, appointed by the government in Ottawa, ostensibly represents the British crown but in the real sense represents the federal government. The lieutenant governor calls upon the recognized leader of the majority party in the unicameral Legislative Assembly to form a government, with that leader as premier. The premier and the Cabinet, selected from the elected members of the legislature, constitute the Executive Council. The Cabinet is responsible to the Legislative Assembly for the operation of the day-to-day business of the government.

Two political parties dominate provincial politics. One, the Social Credit Party, defends the role of the market in economic affairs, favours freeing economic development from the government, and is socially conservative; the other, the New Democratic Party, emphasizes issues of social and economic justice, sees an active role for government in a mixed economy, and is social democratic. The former, a provincial party, held power for most of the time after 1952. In federal elections the New Democrats compete with two national parties, the middle-of-the-road Progressive Conservatives and Liberals.

Regional division

About four-fifths of the population resides in municipalities governed by elected mayors and councils. These local governments are created, supervised, and, to a considerable extent, funded by the provincial government. British Columbia's unorganized territory is divided into 28 regional government districts that provide services to rural areas and have assumed the responsibility for many municipal services.

By agreement, the Royal Canadian Mounted Police (RCMP) act as the provincial police force under the control of the attorney general. Most cities and district municipalities are responsible for law enforcement within their own boundaries. Most elect to have this function contracted out to the RCMP, but Vancouver, Victoria, Nelson, and New Westminster maintain independent forces.

Education. The provincial government spends about one-third of its budget on the rapidly expanding education industry of the province. Public school education is free and compulsory to the age of 15. Postsecondary education is provided by five degree-granting institutions, among which are the University of Victoria in the capital, Simon Fraser University, the University of British Columbia, and the Open Learning Agency in Greater Vancouver, the last specializing in correspondence courses. Technical and vocational instruction is provided by several schools, the most important of which is the British Columbia Institute of Technology in Burnaby.

Health and welfare. The British Columbia Overall Medical Services Plan is available to all residents of the province on a voluntary subscription basis at low, uniform rates. It provides comprehensive medical care coverage, including specified services of physicians, surgeons, certain dental surgeons, the Red Cross, nurses, chiropractors, osteopathic physicians, and others. Hospital care is also available under the British Columbia Hospital Insurance Service (BCHIS). Medical care for the aged meets the standards of the Medical Care Act of Canada and receives contributions from the federal government.

Cultural life. The majority of cultural activities in British Columbia tend to reflect the pioneer background of the province, the great distances between pockets of population, and the various economic backgrounds of areas still oriented to resource industries and to the outdoors. The Williams Lake Stampede is the great annual event of the white and the Indian populations of the ranching country of the Central Interior. Even the annual Pacific National Exhibition in Vancouver is largely oriented to the agricultural communities of the lower Fraser valley rather than to the urban interest of Vancouver. Other centres hold annual Loggers Day festivals, salmon contests, and pioneer events related to outdoor industry and pursuits.

Urban culture

Cultural activities of the larger cities are in a transitional stage, neither frontier nor unabashedly big city. Vancouver's opera association, both respectable and ambitious, mounts productions featuring artists imported from the world's great opera houses, while its symphony society receives enthusiastic support from a loyal following of citizens. Among the most noteworthy cultural constructions in Vancouver are the Centennial Museum, Maritime Museum, and H.R. MacMillan Planetarium, the Bloedel Conservatory, the Queen Elizabeth Theatre complex, the Vancouver Art Gallery, and the Orpheum Theatre. Prince George boasts a fine library that offers innovative programs to people of the Central Interior, while on the south coast Nanaimo and White Rock present excellent summer theatre. Vancouver's Folk Music Festival, held in mid-July at Jericho Beach, is among the most innovative folk festivals in North America. In Victoria the Royal British Columbia Provincial Museum is outstanding.

Local industrialists have done much to support the recent upsurge in cultural activities through their financial support of such distributing agencies as the Vancouver Foundation and the Koerner Foundation. Private funds are channeled through them into the theatres, orchestras, university libraries, art galleries, and museums, as well as to private artists and scholars.

In a special place is the continuing restoration of Barkerville, once the thriving centre of the Cariboo gold rush of the 1860s. There the provincial government has rebuilt scores of old buildings to re-create the old gold town as it appeared more than a century ago. The restoration of historic buildings of Victoria also reflects the province's cultural heritage. Scores of derelict buildings of historic interest have been restored and rented to merchants. Squares have been opened, and whole streets have been revitalized without recourse to bulldozers or slum clearance.

Commercial, recreational, and athletic activities are concentrated in the southwestern part of the province, as are other cultural activities. Spectator sports gained impetus with the building of Empire Stadium in Vancouver in the 1950s and can also now be presented in that city's 60,000-seat domed stadium, BC Place. Vancouver has professional teams in the National Hockey League and the Canadian Football League.

HISTORY

Indian societies

At the time of initial contact with white explorers, Indians in present-day British Columbia numbered about 80,000. The coast was dominated by Coast Salish, Nootka, Kwakiutl, Bella Coola, Tsimshian, and Haida, whose economy was based on the products of the sea and on the huge coastal cedars. Expert fishermen, they utilized traps, nets, hooks, spears, and even an ingenious toggling harpoon for hunting whales. Their clothing was made of skins and cedar bark covered by beautifully patterned blankets woven from the wool of mountain goats. Indian dwellings were large rectangular buildings of cedar beams and planks, divided into compartments for families. Houses were located in clusters along beaches suitable for canoe landings and just above the high-water mark. These Indians were already enterprising traders in copper, blankets, elk hides, furs, shells, candlefish oil, and slaves along the intertribal routes that ran north–south into California and east–west into the interior. They also enjoyed a rich social life in this land of the potlatch. Rival families competed with each other to distribute blankets, food, jewelry, and other favours to guests, often invited from hundreds of miles away, to mark the birth, adolescence, marriage, or death of an important member of the tribe.

Explorations and trading posts. The area that was to become British Columbia first caught the attention of European nations in the late 18th century. Spanish ships visited the coast in 1774, followed by Captain James Cook, who was searching for the Northwest Passage. The latter's account of the fur wealth of the area stimulated the interest of British and American traders, who soon arrived to trade with the Indians for the highly prized sea otter pelts. The growing interest of Great Britain in the area was indicated by the dispatch of Captain George Vancouver, who circumnavigated Vancouver Island and charted the mainland's intricate coastline.

Simultaneously, other British fur traders penetrated the region from the east. Alexander Mackenzie of the North West Company of Montreal entered the region through its winding waterways; he completed the first overland journey across the entire continent when he arrived at Bella Coola, at the mouth of Dean Channel, in 1793. A fur trade based on fixed posts in the interior followed the establishment, by Simon Fraser in 1805, of the first trading post at McLeod Lake. Three years later he descended the Fraser River to its mouth, the site of present-day Vancouver.

Crown colony status

After the southern boundary was fixed at the 49th parallel, Vancouver Island was recognized as solely British territory, following years of near-conflict with the United States. Fort Victoria became the western headquarters of the Hudson's Bay Company. In 1849 Vancouver Island was made a crown colony by the imperial government, which expected that an orderly settlement in this distant outpost of empire would follow. However, the determination of Governor James Douglas to encourage the fur trade and the lure of the California goldfields impeded settlement so successfully that in 1855 the total population of European origin in the colony was only 774, most of them involved in the pursuit of the fur trade.

Gold rush and permanent settlements. The gold strike of 1858 made Fort Victoria into a city, opened the mainland to settlement, and transformed the frontier into a prosperous and dynamic society that was proclaimed the Colony of British Columbia. Hordes of gold seekers from California, Australia, and other parts of the Pacific community joined with British and Canadian migrants to work the alluvial gold deposits of the lower Fraser and by 1862 the Cariboo gold country on the upper Fraser.

By 1865 the gold days were over, and most of the miners had departed with the larger part of nearly $25 million in gold dust. But the rush had attracted an army of ranchers, farmers, hotel operators, storekeepers, and civil servants who, although diminished in numbers, formed a nucleus for an ongoing, settled society. Left behind also was a fairly well-established transportation and communications network. Steamships linked Victoria and points on the Fraser River, and a 400-mile coach road connected Yale, on the lower Fraser, with Barkerville, in the Cariboo gold country. The two colonies joined Canada in 1871 as the Province of British Columbia.

The arrival of the Canadian Pacific Railway at Port Moody in 1885 opened a new era. Permanent railroad and lumbering settlements sprang up along the railroad route. The extension of the line into Vancouver was completed in 1887, the year after the city was incorporated. The establishment of a steamship line connecting it with the Orient, in 1891, assured Vancouver's future as a port.

The early 20th century. Once the construction of the Panama Canal was assured, British Columbia experienced its first boom since the heady days of the gold rush, a boom based upon the assumption of investors that the province's raw resources would soon be able to compete in the markets of the Atlantic nations. Forest tracts were allotted to eager American, British, German, and Canadian investors by a poorly funded provincial government eager to make an impressive fiscal showing and ensure its perpetual reelection. A similar situation prevailed in the mining industry, with the initial prizes going to British investors. Simultaneous with the opening up of these resource industries, the rapid settlement of the treeless prairies occurred, and the demand for lumber was best filled from British Columbia. As the economy revitalized, the population increased from less than 180,000 to nearly 400,000 between 1901 and 1911, and Vancouver emerged as the leading city, with more than 111,000.

Anti-Asian activity

After 1918, postwar dislocation, strikes, and sporadic unemployment fed the flames of anti-Asian feeling, always strong in Vancouver. Politicians, union leaders, and the *Vancouver Daily World* (now defunct) eagerly laid the blame for lack of prosperity on Asian workers. Exclusionist legislation was passed by an obliging provincial legislature in 1923 and remained in force until the late 1940s, when Asians were enfranchised.

Scarcely recovered from the war years, the province found its newly acquired markets swept away by the economic depression of the 1930s. Recovery was delayed until the economy was stimulated once again, by war in 1939. This time, military demands were extensive enough to ensure the continued prosperity of the province's industry even beyond the war years.

Postwar political consolidation. With postwar economic prosperity came political change. To meet the political threat of the depression-born Cooperative Commonwealth Federation (CCF), a moderate socialist party, the Conservatives and Liberals combined forces to form a coalition government during the war years. Growing disenchantment with the party in power made inevitable the appearance of a new party as an alternative to the CCF. The Social Credit Party, which won the election of 1952, was formed by dissidents of both old-line parties but chiefly by conservatives. The Social Credit government remained in power until 1972, when it lost the provincial election to the New Democratic Party (reformist successor to the CCF). In the election of 1975 the Social Credit Party regained control of the Legislative Assembly, and it remained in power through the 1980s. In that decade the boom- and bust-cycles characteristic of most resource-based economies returned, especially affecting the vast region outside of greater Vancouver. (J.C.L./R.A.J.McD.)

NORTHERN TERRITORIES

Northwest Territories

The Northwest Territories constitute more than a third of Canada, itself the world's second largest nation in area, and they reach almost from the eastern to the western extremities of the country across the roof of the North American continent, occupying an area of 1,322,909 square miles (3,426,320 square kilometres). The territories are among the most sparsely populated but habitable regions of the world. In the north the territory extends far above the Arctic Circle to incorporate thousands of islands, the largest of which are Victoria Island in the west and Baffin Island in the east, as well as the islands within bays and straits of Hudson Bay and Ungava Bay. The land is one of high plateaus separating a virtually Arctic lowland on the east from a western depression of more moderate climate. Thousands of lakes dot its heavily glaciated surface.

Administrative regions

There are five administrative regions in the territories. The Baffin region includes the Arctic islands of the north and Melville, Brodeur, and smaller peninsulas. The District of Keewatin region lies north of the province of Manitoba. The Fort Smith region borders the Yukon Territory to the west and Saskatchewan, Alberta, and British Columbia to the south. The Inuvik region is bordered by the Beaufort Sea in the northwest and the Yukon Territory in the southwest. The Kitikmeot region comprises Victoria Island, part of Prince of Wales Island, and the Boothia Peninsula. The capital and largest town is Yellowknife.

PHYSICAL AND HUMAN GEOGRAPHY

The land. Two main types of landscape blend into one another along the timberline, which runs southeast from near the Mackenzie delta on the Arctic Ocean to near Churchill, Man., on Hudson Bay. Southwest of this line lies the northernmost part of the Canadian boreal forest, extending westward to the mountain ranges that separate the Northwest Territories from Yukon Territory. North and east of the timberline stretch the relatively barren grounds of the Arctic, vast reaches of flat, often poorly drained lowlands and of Precambrian rock more than 570 million years old. Within each of these two main regions, the surface vegetation and the animal life it supports vary with soil and climatic conditions.

The Mackenzie Lowlands. The most favourable conditions are found in the Mackenzie Lowlands in the west, where forests of black and white spruce mixed with deciduous species extend north to the Mackenzie delta. The climate here is relatively mild, with warm and dry summers during which mean July temperatures of 60° F (about 16° C) are recorded at most of the settlements along the Mackenzie River. The winters are long and cold, with an average mean January reading at Yellowknife on the northern shore of Great Slave Lake of −18° F (−28° C). With only about 70 frost-free days, the growing season is short. While it lasts, however, wildflowers and grasses flourish, and root and cereal crops can be cultivated. Many species of valuable fur-bearing animals are found, notably muskrat and beaver. Moose, wolves, black and grizzly bears, and mountain sheep and goats are also native to the area. Pickerel, northern pike, and whitefish inhabit the rivers and lakes.

Mackenzie Lowlands permafrost

The cool climate limits navigation on the Mackenzie River system to about four months a year and causes a permanently frozen subsoil, or permafrost, except in a small area south of Great Slave Lake. Permafrost creates serious construction problems, especially where the subsoil is an unstable mixture of fine silt and water.

The northeast. East of the Mackenzie Lowlands and the tree line the climate is colder, and the terrain changes to that of the ancient and rocky Precambrian mass known as the Canadian Shield, the edge of which is straddled by the two largest lakes, Great Bear Lake, 12,095 square miles, and Great Slave Lake, 11,030 square miles. Tree growth becomes sparse and stunted and eventually disappears, to be replaced by the light but tough vegetation of the Arctic tundra. East of the timberline the Arctic climate prevails, with average winter temperatures similar to those in the subarctic region to the west but with summer temperatures not rising above an average of 50° F (10° C) even in July. In these so-called barren lands, annual precipitation is light, and the soils, where they exist at all on the heavily glaciated surface, are usually sandy and thin. Mosses, lichens, and many small hardy flowering plants survive in these conditions and support a variety of animal life ranging from small burrowing mammals and their predator, the white arctic fox, to the large caribou and musk-ox. Seals, walrus, and polar bears are prevalent along the coasts. Bird life is plentiful in summer, with some species, notably the ptarmigans and ravens, remaining all winter. Mosquitoes and other insects abound during the summers.

Settlement patterns. The native people of the territories once led nomadic lives, surviving by adapting to the harsh natural environment. But this balance was disturbed when Europeans established permanent settlements and introduced firearms. The numbers of barren-ground caribou were drastically reduced, and the musk-ox was almost exterminated before being protected in 1927. Attempts to introduce reindeer and other domesticated animals have not been successful.

Nearly all the population lives in small settlements along the Mackenzie River and along the Arctic coastlines of the mainland and northern islands. In addition to Yellowknife, the main towns are Hay River, Fort Smith, and Inuvik; all are in the Mackenzie area.

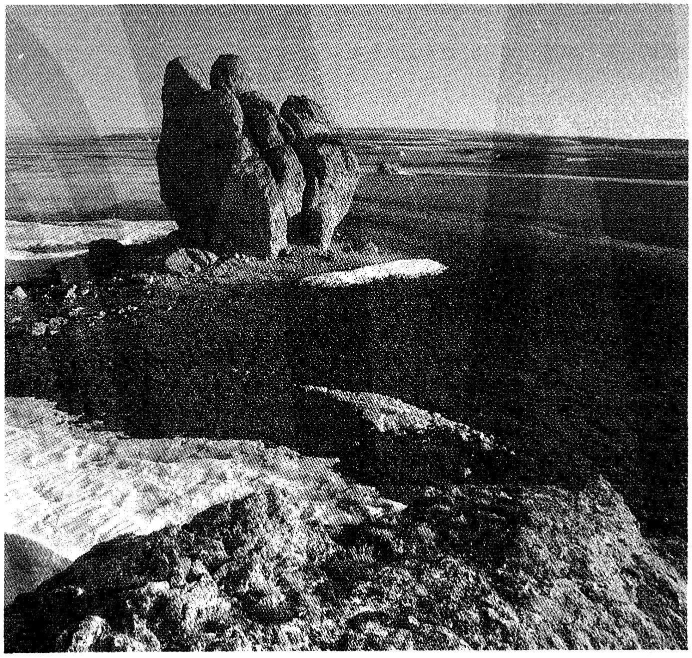
In the eastern Arctic, settlements are very small with the exception of Iqaluit (formerly Frobisher Bay), on Baffin Island between Hudson Bay and the Atlantic, which is the administrative and commercial centre of the region.

The people. The native American Indians and Inuit are two distinct groups, differing in appearance, language, and culture and living apart from one another.

American Indians. The Indians, who make up about one-fifth of the total population, are concentrated in the Mackenzie valley area. These people, who now refer to themselves as the Dene, belong to several tribes, all part of the Athabascan language family. Tribal organization was never strong among the northern Indians, and small bands led by individuals chosen for their skill in the hunt were the effective social unit. This arrangement was easily molded to the needs of the fur trade when it reached the Mackenzie area in the 18th century. Thereafter, the exchange of furs for imported goods became the basis of the Indian economy. Government treaties were made with the groups living south of Great Slave Lake in 1899 and with those living farther north only in 1921. No reservations were established. The decline of the fur trade in the 20th century left many Indians unemployed, and they have had difficulty adapting to wage employment.

The Indian fur-trading economy

Brian Milne/First Light



Rock-strewn tundra of the barren Arctic lands of Polar Bear Pass on Bathurst Island, N.W. Terr.

The Inuit. The origins of the Inuit living in the territories, largely in the coastal areas, are obscure. They constitute about one-third of the territorial population. Although several dialect groups are represented, all apparently have descended from what is known as the Thule culture, a prehistoric maritime society. Early contacts with explorers and whaling crews introduced new diseases and reduced the population during the 19th century. There was also considerable interbreeding. The fur trade was not well established in the Arctic until early in the 20th century; but the Inuit adapted quickly to it, and they, like the Indians, came to depend upon outside sources of supply for most of the necessities of life. Construction activity during World War II and in the postwar years further affected their way of life. Inuit adapted readily to the opportunities for casual employment, and many were quick to abandon the seminomadic trapping and hunting existence for life in the settlements. Canadian government policy in the 1950s and '60s promoted this trend in the interests of upgrading the quality of life.

Europeans. The remaining one-half of the people in the territories are mainly of European and mixed European and native descent. Most live in the more economically advanced Fort Smith region, where they find employment in mining, transportation, and public service. Much of this population has always been transient.

The economy. High production costs and transportation problems inhibit development of many of the territories' mineral resources, including the petroleum and natural gas fields that exist in the western Arctic coastal regions.

Government assistance in the development of major resources has been provided mainly in the form of roads, electric-power facilities, mapping, and geologic services, although the government has participated in the search for oil and natural gas. Government agencies produce and distribute electric power throughout the territories and provide certain transportation services.

Resources. Mining has been the principal nonrenewable resource industry of the territories. Gold has been mined at Yellowknife on the north shore of Great Slave Lake since the late 1930s. Radioactive ores were mined at Port Radium on Great Bear Lake in the 1930s, initially to produce radium; later, during World War II and afterward, uranium was mined. Large-scale lead and zinc mining was carried on at Pine Point from the late 1960s until the 1980s, when the ore deposit became depleted. Smaller metal mining projects have been carried out at widely scattered sites, a few in the high Arctic regions. Silver, copper, tungsten, and cadmium are among the metals that have been produced. Petroleum fuels for use in the territories are obtained from refineries located at the Norman Wells and Pointed Mountain fields. Gasoline and diesel fuel are important both for transportation and for electric power generating purposes. Large-scale hydroelectric power development has not been feasible, but a number of small sites have been developed.

Hunting and fishing. Trapping continues to provide income for some of the native population. Muskrat, beaver, marten, mink, and lynx are the most important furs taken in the Mackenzie area, while the white fox remains the principal fur in the Arctic regions. Fishing and hunting of sea mammals also provides some employment. Whitefish and lake trout are fished commercially on Great Slave and some smaller lakes. Arctic char are fished and exported to southern markets from several communities in the eastern Arctic. Seals and small whales are hunted for food, and some sealskins are marketed commercially. Sport fishing and hunting are major attractions for the small but growing number of tourists who visit each year.

Forestry and agriculture. The Fort Smith region has most of the 200,000 square miles of forested land, but even there, large stands of merchantable timber are not plentiful. The several sawmills process the timber only for local use. Although there are more than one million acres of arable land, farming is not profitable. Some field crops are grown for local use, but most foodstuffs must be imported from the outside, greatly affecting their price.

Airline services

Transportation. Nearly all passenger and much freight traffic is carried by air services. Flights link Yellowknife and other major settlements along the Mackenzie valley to Edmonton, Alta. Settlements on the west coast of Hudson Bay and in Keewatin region are connected to Winnipeg, and those in the eastern Arctic to Montreal.

Surface transportation for heavy freight is mainly by water. Fuel oil for heating, and other bulky supplies, are carried to eastern Arctic settlements by seagoing supply ships organized by the federal Department of Transport. Government departments also supply remote villages and military installations along the eastern coastline. The Northern Transportation Company Limited is the principal carrier on the Mackenzie River waterway. The waterway is supplemented in the southern part of the Fort Smith region by the Mackenzie Highway and a railway connecting Pine Point to the trans-Canada rail systems at Grimshaw, Alta. The southern Mackenzie valley is linked by a highway from Fort Simpson to the Alaska highway system in Yukon, while to the north the Mackenzie delta is connected to Dawson, in Yukon, by the Dempster Highway. Tractor trains and other overland vehicles using temporary winter roadways carry freight into remote areas. Snowmobiles are used for light winter travel.

The commissioner and assembly

Administration and social conditions. *Government.* Ultimate constitutional responsibility for government in the territories rests with the federal government in Ottawa, but most provincial responsibilities have been delegated to a territorial administration sitting at Yellowknife. This consists of a commissioner and a Legislative Assembly comprising 24 members elected by the people to serve a four-year term. The elected council reflects the distinct ethnic mix of the territorial population and has come to be composed of roughly equal numbers of Inuit, Indians, and whites. The council generally convenes during the winter, spring, and fall. During these sessions the council has the authority to enact legislation on most territorial matters. All legislation is subject to the commissioner's assent and may be revoked by the federal government within one year. An executive committee advises the commissioner on policy matters. The commissioner, in turn, proposes legislation to the council, but council members may introduce private bills of a nonfiscal nature. Because royalties and other revenues from natural resource use in the territories are collected by the federal government, the territorial administration relies for most of its revenues on funds transferred to it from the federal authority.

The territories are represented by one elected member of Parliament. Justice is dispensed by a territorial court, a police magistrate, and several justices of the peace. Law enforcement is carried out by the Royal Canadian Mounted Police. The federal government controls all natural resources except game and administers them through its Department of Indian Affairs and Northern Development.

Education and social services. Missionaries provided nearly all the education and health care available in the territories until the 1950s, but since then both have become mainly government responsibilities. The federal Department of Indian Affairs and Northern Development cooperates with the territorial department of education in providing elementary and secondary schooling and in assisting native students to pursue postsecondary education outside the territories. A number of postsecondary programs and courses are offered by a community college system at several centres throughout the territories. Health care is provided through comprehensive territorial hospital and medical-services insurance plans.

Cultural life. Modern forms of transportation and communication have done much to break down the isolation of life in the north, and contemporary North American popular culture is evident in most communities.

Inuit culture

Satellite television has made a wide range of entertainment and educational programs available to viewers in even the most remote settlements. Radio stations relay programs throughout the territories, and most of the larger settlements have their own weekly newspapers. Some even have local television stations that originate programs for distribution in the territories. In the Fort Smith region, little remains of the traditional native culture, although there is growing interest, particularly at the local level, in revitalizing the linguistic and cultural heritage of the Dene. In the eastern regions, elements of the Inuit culture have been preserved. Public policy in recent years has encouraged development of Inuit traditions in arts and crafts, providing an important source of income in some Inuit communities and making Inuit culture familiar to collectors worldwide. Inuit-language broadcasting is well established, and a written language has been introduced to encourage Inuit writing and to facilitate communication among the communities.

HISTORY

Vikings probably visited parts of the Canadian Arctic during the Middle Ages, but there are no records of exploration until the voyage in 1576 of the English mariner Martin Frobisher in search of a northwest passage to the Orient. Other expeditions in the 17th century also failed to find the route, but they added to knowledge of the Arctic regions. Interest in finding the route waned in the 18th century, but whaling ships became commonplace in the Arctic waters. The first recorded exploration of the mainland was by Samuel Hearne, who in 1770–72 journeyed from the west coast of Hudson Bay to the mouth of the Coppermine River on the northern coast. Other inland explorations were mainly the work of Montreal-based fur traders. In 1789 Alexander Mackenzie of the North West Company traveled down the river that bears his name to reach the Arctic Ocean. In the 19th century there was renewed interest in a northwest passage. Sir John Franklin and others explored much of Mackenzie District (now largely the regions of Fort Smith, Kitikmeot, and Inuvik) and mapped parts of the northern coastline during the 1820s, work that Thomas Simpson continued in 1838–39. Searchers for the lost Franklin expedition of 1845–48 explored and mapped other parts of the eastern Arctic in the following decade. Later a series of expeditions attempted to reach the North Pole; such exploits continued into the 20th century but by then were overshadowed by more practical activities directed at identifying the resource potential of the Canadian North.

Early administration

Settlements were first established to serve the whaling fleets and fur traders. Missionaries became active in the Mackenzie valley in 1852 and in the eastern Arctic toward the end of the century. No resident administrative authorities were established within the present limits of the Northwest Territories until the 20th century. Responsibility for the mainland territories that drain into Hudson Bay, known as Rupert's Land, was vested in the Hudson's Bay Company. The remaining part of the mainland, the North-Western Territory, was under nominal British rule until 1870, at which time both it and Rupert's Land were ceded to Canada. In 1880 the Arctic islands claimed by Britain were also placed under Canadian jurisdiction. Separation of the Yukon Territory, creation of new provinces, and enlargement of other provinces reduced the Northwest Territories to its present limits by 1912. The Royal Canadian Mounted Police were made responsible for maintaining law and order and for providing whatever governmental administration was required in the area.

Fur traders, missionaries, and the police directed the life of the Northwest Territories until the 1920s, when discovery of oil near Fort Norman on the Mackenzie River prompted the Canadian government to establish a territorial administration for the area. Mining replaced the fur trade as the most important industry in Mackenzie District in the 1930s. World War II brought much government-financed construction activity to the territories. In the southern Mackenzie area the Canol pipeline, linking the oil field at Norman Wells to a refinery at Whitehorse in the Yukon, and construction of several large airfields in the eastern Arctic did much to open the Canadian North to further exploration and development. After the war, construction of the Mackenzie Highway to Great Slave Lake and the building of the Distant Early Warning radar network, the DEW line, continued this process. A great expansion of government-sponsored health, education, and welfare services transformed living and social conditions throughout the North.

Yukon Territory

A region of rugged mountains and high plateaus, the Yukon Territory comprises 186,661 square miles (483,450 square kilometres) in extreme northwestern Canada. Its mineral wealth has long been known, but the combination of an Arctic climate with remoteness from markets has minimized the economic exploitation of such resources and the development of modern settlement. Instead, the territory remains among the few frontiers on the North American continent, a sparsely populated and largely unspoiled wilderness.

The Yukon Territory lies directly north of the province of British Columbia, with the Northwest Territories to the east and the U.S. state of Alaska to the west. In the north it stretches well above the Arctic Circle to border the Beaufort Sea. The capital is Whitehorse.

PHYSICAL AND HUMAN GEOGRAPHY

The land. The territory lies within the mountainous cordilleran region of western North America. The more settled areas lie in a large central plateau surrounded by mountains and drained by the Yukon River system flowing northwestward into Alaska. Some of the surrounding mountains are spectacular, especially the St. Elias Mountains in the southwest, which have some of the highest peaks in North America, including Mount Logan at 19,524 feet (5,951 metres).

The northern stretches of flat and poorly drained tundra have widespread permafrost, a permanently frozen ground that makes construction of most kinds difficult. Although temperatures vary, sometimes reaching summer highs of 95° F (35° C) and winter lows below −60° F (−51° C), the monthly average readings are not unlike those found in some Canadian cities farther east and south. Summers are short but the days are long, especially in the northern part of the territory lying above the Arctic Circle. Precipitation is light, averaging only 10 inches (250 millimetres) annually at Whitehorse.

Environmental impact of people

Vegetation is sparse in many regions because of the dry, cool climate and poor soils, but some of the southern valleys are heavily forested. Animal life, on the other hand, abounds in the Yukon. Large game animals and smaller fur-bearing mammals are native to the area, as are waterfowl and other common North American birds. Common species of fish include the arctic grayling, lake trout, and northern pike. Much of the territory remains unspoiled wilderness, but the impact of people on the environment is apparent. Near the settlements, reliance on wood for fuel has destroyed timber, and in more remote places forest fires caused by humans have often gone uncontrolled. In addition, surface mining has scarred the landscape near Dawson and Whitehorse.

The city of Dawson and some older Indian villages such as Old Crow, situated well above the Arctic Circle, are set in locations of great natural beauty and retain a picturesque appearance. More modern centres of mining and transportation, such as Whitehorse, Watson Lake, and Mayo, resemble towns of like size elsewhere in western North America. On the fringes of some of these, rude settlements provide shelter for people forced from their more remote habitations by the decline of trapping as an economically viable activity and attracted by the social services available in the more urban areas. Some two-thirds of the total population of the Yukon live in Whitehorse.

The people. About 15 percent of the people are native American Indians; most belong to the Athabascan language family, of which the Kutchin group is the largest. Although registered Indians in the territory are a federal responsibility, there are few occupied Indian reservations. In 1973 the Canadian government recognized the validity of Indian territorial claims based on past treaties, thereby rendering the first official recognition of such rights. The rest of the population comprises people of European descent or of mixed Indian and European stock.

Health and welfare services have been greatly improved in recent decades, contributing to a decline in the total mortality rate. A high infant mortality remains a serious problem, but fertility rates are above the Canadian average, giving the territory a high rate of natural increase in population. The ability of the economy to absorb this increase has become a matter of increasing concern.

Few people continue to make their living by hunting, fishing, or trapping, and, except for some isolated groups, most of the people are now dependent upon wage employment. Wages and salaries are unusually high in most occupations. Since nearly all food, fuel, building materials, and manufactured goods are imported over considerable distances, the cost of living is also high. Marked differences in living standards exist between those who find employment in mining, transportation, and public service and those who lack the education required for such employment.

Value of mining

The economy. Since the late 19th century, mining has been the Yukon's most important industry, although market fluctuations can cause sharp variations in production. The extensive and varied mineral resources include large deposits of combined silver-lead and lead-zinc ores and of copper, coal, iron, and some petroleum and natural gas. Commercial production of silver-lead-zinc and of copper concentrates has been accomplished, but the economic feasibility of developing other known mineral deposits is not established.

Forest resources are limited, but commercial logging is carried out in some southern areas. Small sawmills produce some of the lumber required for local building, for mining timbers, and for fuel, although imported fuel oil is replacing both wood and locally mined coal as the principal fuel. Farming is possible, but climate and soil conditions make it so unprofitable that virtually no settlement is undertaken for agricultural purposes. Trapping is still important despite its decline, and fox, beaver, wolf, and lynx are caught for their pelts.

Tourism is a rapidly expanding industry, however, now ranking second to mining in the value it contributes to the economy. Both federal and territorial governments play an important role in the economy. Since the end of World War II, government has assumed major responsibility for

George Hunter

Dawson, on the bank of the Yukon River in the Yukon Territory.

building roads, airfields, and electric power facilities, as well as for providing incentives to private companies interested in developing resources in the area. Exploration for oil and gas began in the 1950s. Commercially significant gas reserves were discovered in the 1960s, and there was subsequently extensive exploration throughout the territory for petroleum resources.

Modern transportation facilities link the Yukon to the outside world. Scheduled jet air services operate between Whitehorse and Edmonton, Alta., and Vancouver, B.C. Until the White Pass and Yukon railway closed in 1982, it carried freight and passengers between Whitehorse and Skagway, Alaska, from which point ocean services to other west coast ports were available. The Alaska Highway connects the territory to the North American highway system and serves as the backbone for a limited network of roads within the territory. In the late 1970s an important highway link was provided between a point on the Klondike Highway near Dawson to Inuvik in the Mackenzie delta (the Dempster Highway). Helicopters and light aircraft are the normal means of travel to more remote areas. The Whitehorse airport is the most important in the Yukon.

Regional government

Administration and social conditions. With the exception of natural resources, which remain a federal responsibility, most administrative and legislative matters normally under the jurisdiction of Canadian provinces are dealt with by the territorial administration. An appointed resident commissioner oversees the interests of the federal government, but the day-to-day governmental operations are the responsibility of the Executive Council. The council is composed of five members selected from the Legislative Assembly: the elected leader (or the government leader) of the majority party in the Legislative Assembly, and four others nominated by the government leader and appointed by the commissioner. Each member of the Executive Council acts as a Cabinet minister responsible for a specific department of the territorial government. The members of the Legislative Assembly are elected to four-year terms, and the membership of the assembly may be expanded to 25 representatives. Residents of the territory are represented in the federal Parliament by one member of the House of Commons.

Educational, health, and welfare services are administered by the territorial government with financial assistance from the federal government. Primary and secondary education are provided by a territorial school system, and Yukon College, with its main campus at Whitehorse and a network of community branches, provides two years of university-level courses and a number of vocational and adult-education programs. Comprehensive medical and hospital services are provided on the same publicly financed basis as elsewhere in Canada, with the major hospital facilities located in Whitehorse and a system of nursing stations serving more remote centres. Police services are provided by the Royal Canadian Mounted Police.

Reminders of the gold-rush era

Cultural life. Cultural life is dominated by modern media. Satellite communication has made North American television available throughout the territory. Publicly operated radio stations serve the region, and Whitehorse and some communities have local newspapers. There has been a revival of interest in native cultural activities. Promotions keep alive the romanticized image of the Klondike gold-rush era as embodied in the poems of the English-born writer Robert Service, who lived in the Yukon, and more recent writers. Reminders of the gold-rush days are preserved in museums and displays in Whitehorse and Dawson. Some historical sites, notably the Palace Grand Theatre in Dawson, have been restored.

HISTORY

The Yukon was among the last areas of North America to be explored by nonnatives. Two explorers for the Hudson's Bay Company first entered the region around 1840. John Bell came by way of the Peel River from the north, Robert Campbell by the difficult Liard River route from British Columbia. The company subsequently built trading posts in the south, but Indian hostility soon forced them to be abandoned. Farther north, Fort Yukon (in contemporary Alaska) was established in 1847 on the Yukon River in what was then Russian territory. Relocated after the U.S. purchase of Alaska from Russia in 1867, and again in 1890, Fort Yukon remained a centre for a small fur trade. During the early 1870s, gold discoveries on tributaries of the Yukon River attracted prospectors to the area.

Only small amounts of gold had been mined before Aug. 17, 1896, when a pioneer and his two Indian companions found rich deposits in Bonanza Creek, a small tributary of the Klondike River near its confluence with the Yukon. The discovery led to the great gold rush of 1898, at the peak of which the nearby settlement of Dawson grew into a city of some 25,000. Access to the area was quickly improved by construction of a 110-mile (177-kilometre) narrow-gauge railway, the White Pass and Yukon Route, extending from the port of Skagway, Alaska, to Whitehorse on the upper reaches of the Yukon River. In 1898 the Canadian Parliament separated the rapidly growing area from the Northwest Territories and gave it separate territorial status.

The Klondike boom lasted only a few years. By 1900 many of the individual miners had given up and were leaving the Yukon. The gold-mining industry was subsequently reorganized by companies that brought in large-scale mechanical mining techniques. The population fell from 27,219 in 1901 to 8,512 in 1911, and finally to a low of 4,157 in 1921. Although no new gold deposits were located, other minerals were discovered. Copper ore was mined near Whitehorse during World War I, and during the 1920s and 1930s ores containing silver and lead were mined in the Mayo district in the central plateau. Development of these and other resources, however, was hindered by high operating and transportation costs, and in 1941 only 4,914 persons were counted in the census.

Economic expansion during World War II

During World War II such military projects as the Alaska Highway, a part of the Pan-American Highway system, brought a second boom to the territory. Improvements in transportation and communications helped open the territory to greater exploration and development, while increased public interest in the area prompted a great expansion of government services and an increased influx of tourists. Most of this activity centred on Whitehorse, which replaced Dawson as the territorial capital in 1953. Despite such diversification, however, the economy has remained heavily dependent on the mining industry and consequently has been highly unstable. For several years in the mid-1980s, for example, poor market conditions caused every producing mine in the territory to close.

(K.J.R.)

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Northwest Territories and Yukon. MORRIS ZASLOW, *The Opening of the Canadian North, 1870–1914* (1971), and *The Northward Expansion of Canada, 1914–1967* (1988), provide a comprehensive general history of the territories; KEN COATES, *Canada's Colonies: A History of the Yukon and Northwest Territories* (1985), includes a critical analysis of government policies. Economic history is treated in K.J. REA, *The Political Economy of the Canadian North: An Interpretation of the Course of Development in the Northern Territories of Canada to the Early 1960s* (1968). KEN COATES and WILLIAM R. MORRISON, *Land of the Midnight Sun: A History of the Yukon* (1988), describes the gold rush. The early whaling industry is treated in W. GILLIES ROSS, *Whaling and Eskimos, Hudson Bay, 1860–1915* (1975). DIAMOND JENNESS, *Eskimo Administration II: Canada* (1964), remains a standard historical account of government policy toward the native population. SHELAGH D. GRANT, *Sovereignty or Security: Government Policy in the Canadian North, 1936–1950* (1988), examines important changes brought about by World War II. The political development of the territories is treated in GURSTON DACKS, *A Choice of Futures: Politics in the Canadian North* (1981); and WILLIAM R. MORRISON, *A Survey of the History and Claims of the Native Peoples of Northern Canada* (1983). Concerns with environmental protection and the impact of economic development is studied in HUGH BRODY, *The People's Land: Eskimos and Whites in the Eastern Arctic* (1975); and THOMAS R. BERGER, *Northern Frontier, Northern Homeland: The Report of the Mackenzie Valley Pipeline Inquiry*, rev. ed. (1988).

(R.R.K./D.J.Be./S.E.P./L.H./B.O.G./P.Ga./
D.A.E.S./R.B.D./K.S.C./N.W./R.A.J.McD./K.J.R.)

Canadian Literature

Reflective of the country's dual origin and linguistic distribution, the literature of Canada falls into two divisions: English and French. This article provides a brief historical account of each of these literatures.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 621, and the *Index.* (Ed.)

The article is divided into the following sections:

CANADIAN LITERATURE IN ENGLISH

Prose and poetry. *Beginnings, 1600–1900.* The first writers of English in Canada were visitors—explorers, travelers, British officers and their wives—who recorded their impressions of British North America in charts, diaries, journals, and letters. With these journeys and the resulting documents of discovery, the Canadian tradition of embarking upon the search for both a myth of origins and a definition of "Canada" was initiated.

The earliest such documents were unadorned narratives of travel and exploration. The explorer Samuel Hearne wrote *A Journey from Prince of Wales's Fort in Hudson's Bay, to the Northern Ocean* (1795), and Sir Alexander Mackenzie, an explorer and fur trader, recorded his travels in *Voyages from Montreal . . . Through the Continent of North America, to the Frozen and Pacific Oceans* (1801). These accounts, written in plain language, document heroic journeys to the vast, unknown west and north on behalf of the Hudson's Bay Company and the North West Company, the great fur-trading monopolies.

Frances Brooke, the wife of a visiting British military chaplain in the conquered French garrison of Quebec, wrote the first published novel with a Canadian setting. Her *History of Emily Montague* (1769) is an epistolary romance describing the sparkling winter scenery of Quebec and the life and manners of its residents.

Halifax, in the colony of Nova Scotia, and New Brunswick's Fredericton were the scenes of the earliest literary flowering in Canada. The first literary journal, the *Nova-Scotia Magazine,* was published in Halifax in 1789. The town's literary activity was invigorated by an influx of United Empire Loyalists during the American Revolution and by the energetic Joseph Howe, a journalist, a poet, and the first premier of Nova Scotia. Two of the most potent influences on literary development were in evidence by the end of the 18th century: literary magazines and presses, and a strong sense of regionalism. By satirizing the dialect, habits, and foibles of Nova Scotians, or Bluenoses, Thomas McCulloch, in his serialized *Letters of Mephibosheth Stepsure* (1821–22), and Thomas Chandler Haliburton, in *The Clockmaker* (1835–36), featuring the brash Yankee peddler Sam Slick, adroitly brought their region to life and helped found the genre of folk humour.

Early poetry

Most of the earliest poems were patriotic songs and hymns (*The Loyal Verses of Joseph Stansbury and Doctor Jonathan Odell,* 1860) or topographical narratives, reflecting the first visitors' concern with discovering and naming the new land and its inhabitants. In *The Rising Village* (1825), native-born Oliver Goldsmith used heroic couplets to celebrate pioneer life and the growth of Nova Scotia, which, in his words, promised to be "the wonder of the Western Skies." His optimistic tones were a direct response to the melancholy poem by his Anglo-Irish granduncle, Oliver Goldsmith, whose *Deserted Village* (1770) concludes with the forced emigration of dispossessed villagers.

Immigrants, dreaming of a new Eden but encountering instead the realities of unpredictable Indians, a fierce climate, unfamiliar wildlife, and physical and cultural deprivation, were the subject of prose sketches by the Strickland sisters, Susanna Strickland Moodie and Catherine Parr Strickland Traill. Susanna Moodie's harsh, yet at times comical, *Roughing It in the Bush* (1852) was written to discourage prospective emigrants; but Catherine Parr Traill's *Backwoods of Canada* (1836) presents a more favourable picture of the New World.

The Dominion of Canada, created in 1867 by the confederation of Nova Scotia, New Brunswick, Upper Canada, and Lower Canada (now Ontario and Quebec), precipitated a flurry of patriotic and literary activity. The so-called Confederation poets turned to the landscape in their search for a truly native verse. Unlike their predecessors, they no longer merely described or moralized nature but attempted as well to capture what Archibald Lampman called the "answering harmony between the soul of the poet and the spirit and mystery of nature." The New Brunswick poet Charles (later Sir Charles) G.D. Roberts inspired his cousin, the prolific and vagabond Bliss Carman, and the Ottawa poets Lampman and Duncan Campbell Scott to begin writing verse. Lampman is known for his meditations on the landscape, and Scott is remembered for his sympathetic presentation of Indians trapped by the white man's world. Perhaps the most original poet of this period was Isabella Valancy Crawford, whose colourful mythopoeic verse, with its images drawn from Indian lore, pioneer life, mythology, and a symbolic animated nature, was published as *Old Spookses' Pass, Malcolm's Katie, and Other Poems* in 1884.

During this period, the historical romance was the most popular form of novel. In his Gothic tale *The Golden Dog: A Legend of Quebec* (1877) William Kirby romanticized the seigneurial life of New France, while John Richardson portrayed the uprising led by the Indian chief of the Ottawas, Pontiac, in *Wacousta; or The Prophecy, a Tale of the Canadas* (1832). More original are James De Mille's satiric travel fantasy *A Strange Manuscript Found in a Copper Cylinder* (1888) and Roberts' quasi-documentary animal stories *The Kindred of the Wild* (1902).

20th-century poetry

Modern period, 1900–60. In the early 20th century popular poets responding to the interest in local colour depicted French-Canadian customs and dialect (W.H. Drummond, *The Habitant and Other French-Canadian Poems,* 1897), the Mohawk tribe and rituals (E. Pauline Johnson, *Flint and Feather,* 1912), and the freedom and romance of the north (Robert Service, *Songs of a Sourdough,* 1907). John McRae's account of World War I, "In Flanders Fields" (1915), remains Canada's best known poem. Slowly a reaction against sentimental, patriotic, and derivative Victorian verse set in. E.J. Pratt created a distinctive style both in lyric poems of seabound Newfoundland life (*Newfoundland Verse,* 1923) and in the epic narratives *The Titanic* (1935), *Brébeuf and His Brethren* (1940), and *Towards the Last Spike* (1952), which through their reliance on accurate detail participate in the documentary tradition. Influenced by Pratt was another innovative and experimental poet, Earle Birney, best known for his tragic narrative "David" (1942).

Toronto's *Canadian Forum* (founded in 1920) and Montreal's *McGill Fortnightly Review* (1925–27) provided an outlet for the "new poetry" and the emergence of modernism. Here and in their anthology, *New Provinces* (1936), A.J.M. Smith, F.R. Scott, and A.M. Klein began their long literary careers. Emphasizing concrete images, open language, and free verse, these modernists felt that the poet's task was to identify, name, and take possession of the land. Klein wrote in "Portrait of the Poet as Landscape" (1948) that the poet is "the nth Adam taking a green inventory/in a world but scarcely uttered, naming, praising." The bonds of a colonial frame of mind characterized by fear of the unknown, reliance on convention, a puritan consciousness—what the critic Northrop Frye, in the "Conclusion" written for the first edition of the *Literary History of Canada* (1965), called the "garrison mentality"—were being broken and cast off.

Strong reaction to the Depression, the rise of fascism, and World War II dominated the poems of the 1930s and '40s. Using the documentary mode, Dorothy Livesay condemned the exploitation of workers in *Day and Night* (1944), while her lyric poems spoke frankly of sexual love (*Signpost,* 1932). In opposition to the cosmopolitan and metaphysical verse promoted by Smith and the literary magazine *Preview* (1942–45), Irving Layton, Louis Dudek, and Raymond Souster urged poets to focus on realism and the native or North American context.

The novel of social realism

By 1900 novels of local colour were beginning to overshadow historical romances. Lucy Maud Montgomery's beloved children's book *Anne of Green Gables* (1908) and its sequels were set in Prince Edward Island. Ontario towns and their "garrison mentality" provided the setting for Sara Jeannette Duncan's portrayal of political life in *The Imperialist* (1904), Ralph Connor's *The Man from Glengarry* (1901), Stephen Leacock's satirical stories *Sunshine Sketches of a Little Town* (1912), and Mazo de la Roche's best-selling Jalna series (1927–60). Out of the Prairies emerged the novel of social realism, which documented the small, often narrow-minded farming communities pitted against an implacable nature. Martha Ostenso's *Wild Geese* (1925), a tale of a strong young girl in thrall to a cruel father, and Frederick Philip Grove's *Settlers of the Marsh* (1925) and *Fruits of the Earth* (1933), depicting man's struggle for mastery of himself and his land, are moving testaments to the courage of farmers.

A tentativeness in form and subject matter pervades the novels published during the 1940s and '50s and is reflected in their protagonists, most of whom are sensitive, restless children or artists. In this category fall the Prairie novels *As for Me and My House* (1941) by Sinclair Ross, *Who Has Seen the Wind* (1947) by W.O. Mitchell, and *The Mountain and the Valley* (1952) by Ernest Buckler, set in Nova Scotia's Annapolis Valley. These novels strain the bonds of conventional narrative structures, as they shift from social realism toward lyricism. In the panoramic *Two Solitudes* (1945) and *The Watch That Ends the Night* (1959), Hugh MacLennan attempted to capture moral, social, and religious conflicts that rend individuals, families, and the two aspects of Canada, French and English. Sheila Watson's enigmatic and mythic *The Double Hook* (1959) and Ethel Wilson's *Swamp Angel* (1954), about a housewife's desertion, present quest journeys against the striking backdrop of British Columbia's interior.

1960 and beyond. After the 1950s this tentativeness in fiction either became itself the subject of the novel or dissipated in more confident forms of writing. Robertson Davies' popular Deptford trilogy (*Fifth Business,* 1970; *The Manticore,* 1972; *World of Wonders,* 1975) examines the growth of its protagonists into maturity within a Jungian paradigm. Alice Munro in *Lives of Girls and Women* (1971), set in southwestern Ontario, and Margaret Laurence in her Manawaka novels (*The Stone Angel,* 1964; *A Jest of God,* 1966; *The Diviners,* 1974) explored their heroines' rebellion against a constricting small-town heritage. In an intricately woven style, Mavis Gallant's stories (*Home Truths,* 1981) depict isolated characters whose fragile worlds of illusion are shattered. With trenchant irony, Margaret Atwood dissected contemporary urban life and sexual politics in *The Edible Woman* (1969), *Surfacing* (1972), *Lady Oracle* (1976), *Bodily Harm* (1981), and *The Handmaid's Tale* (1985). Mordecai Richler's Montreal novels, *The Apprenticeship of Duddy Kravitz* (1959) and *St. Urbain's Horseman* (1971), satirize the condition and hypocrisy of modern society through Jewish black humour, while Leonard Cohen's *Beautiful Losers* (1966) explores the relationship between sainthood, eroticism, and creation.

Many writers published in the 1970s and '80s purposely subverted the conventions of fiction, shifting from realist to surrealist, self-reflexive, feminist, or parodic modes. Although historical events and the investigation of place as an imaginative source remained the most common subject matters, the narrative forms were experimental and playful. Jack Hodgins mapped a surreal island world in *The Invention of the World* (1977), while Audrey Thomas revealed the difficulty of women's lives in innovative short stories (*Real Mothers,* 1981) and novels (*Latakia,* 1979, and *Intertidal Life,* 1984). In *The Studhorse Man* (1969) and *Badlands* (1975), Robert Kroetsch transformed the realism of Prairie fiction into postmodern parodies of the quest journey. In *The Temptations of Big Bear* (1973) and *The Scorched-Wood People* (1977), Rudy Wiebe constructed fictional epics based on historical events in the west. Joy Kogawa's *Obasan* (1981) is a skillful "docufiction" describing the internment of Japanese-Canadians during World War II. George Bowering's *Burning Water* (1980), which focuses on the 18th-century explorer George Vancouver, and Michael Ondaatje's *Coming Through Slaughter* (1976), the story of the jazz musician Buddy Bolden, mingle history with autobiography in self-reflexive narratives that enact the process of writing. In *The Wars* (1977) Timothy Findley's narrator, through letters, clippings, and photographs, re-creates the effects of World War I on his hero.

The blossoming of poetry

In the 1960s and '70s poetry also blossomed, fueled by a resurgence of nationalism, as Canadians asserted their difference from Americans, and by the emergence of literary presses and small literary magazines like *Canadian Literature* (founded in 1959) and *Tish* (1961–69). Atwood's *The Circle Game,* a book of ironic, laconic poems, was published in 1966 and was followed later by *Journals of Susanna Moodie* (1970), in which Atwood translated the 19th-century author of *Roughing It in the Bush* into a modern figure of alienation. Ondaatje also turned to historical persona in his collage *The Collected Works of Billy the Kid* (1970), as did Bowering in his long poem "George, Vancouver" (1970). Daphne Marlatt reinvented a town's past in *Steveston* (1974) and Kroetsch his Prairie roots in *Field Notes* (1981). Fascination with place and history also permeates Al Purdy's poems about the country north of Belleville, Ont., and about his travels west and to the Arctic (*Being Alive,* 1978). Margaret Avison (*Winter Sun/The Dumbfounding,* 1982), Anne Wilkinson (*Collected Poems,* 1968), P.K. Page (*Cry Ararat,* 1967), Eli Mandel (*Dreaming Backwards,* 1981), Phyllis Webb (*The Vision Tree,* 1982), and D.G. Jones (*A Throw of Particles,* 1983) grappled with metaphysical concerns and the fragile particles of existence through images drawn from places, travel, and mythology. Regions, particularly the Prairies, grew stronger as centres of imaginative focus, while the desire of women to express their distinctive voices and experiences resulted in stimulating literary journals (*Room of One's Own; Atlantis*) and anthologies (*Writing Right,* 1982; *Stories by Canadian Women,* 1984).

Drama and criticism. Like the poets and novelists, Canadian dramatists in their quest for a myth of origins have often turned to historical incidents. The earliest forms of dramatic writing, Charles Mair's *Tecumseh* (1886) and Sarah Anne Curzon's *Laura Secord, the Heroine of 1812* (1887), both based on the War of 1812, were in verse. In the 1920s and '30s Merrill Denison, Gwen Pharis Ringwood, and Herman Voaden struggled to establish Canadian drama, relying on the amateur little theatres for support. By the 1950s and '60s several professional theatres had opened, producing a more sophisticated milieu for dramatists like John Coulter, whose *Riel* (1962) creates a heroic figure of Louis Riel, the leader of the Métis rebellion. As regional and experimental theatres multiplied, increasingly innovative and daring productions were mounted: examples were John Herbert's *Fortune and Men's Eyes* (1967), on homosexuality in prison; George Ryga's *The Ecstasy of Rita Joe* (1971), about an Indian prostitute; and James Reaney's Donnelly trilogy (1976–77), about the feuds and the massacre of an Irish immigrant family in southern Ontario. Collective creations resulted in *The Farm Show* (1976), *Paper Wheat* (1978), and Rick Salutin's *1837* (1976) and *Les Canadiens* (1977), all flavoured by a strong sense of regionalism. Stark realism shaped David Freeman's *Creeps* (1972), David French's *Leaving Home* (1977), David Fennario's *On the Job* (1976), and Michael Cook's *The Head, Guts and Sound Bone Dance* (1974). Women's lives in the past were the focus of Carol Bolt's *Red Emma* (1974); Sharon Pollock's *Blood Relations* (1981), a powerful drama about the accused murderer Lizzie Borden; and Betty Lambert's *Jennie's Story* (1984). In *Billy Bishop Goes to War* (1981) John Gray created a musical from the story of a well-known World War I fighter pilot.

Literary criticism

Critics have tended to discuss Canadian literature by outlining the stages of its literary evolution from the colonial beginnings through an immigrant and derivative period to the flowering of a national literature. By the 1920s there were several descriptive and evaluative surveys, such as Archibald MacMechan's *Head-waters of Canadian Literature* (1924). During the modern period, criticism dealt primarily with poetry, analyzing its origins, mythmaking tendencies, and degree of sympathy with nature. With the publication of the *Literary History of Canada* in 1965, Canadian literature developed into a field of serious study. Frye's "Conclusion," which describes how the Canadian imagination embarks on a search for identity, influenced Atwood's *Survival* (1972) and D.G. Jones's *Butterfly on Rock* (1970). Frank Davey's article "Surviving the Paraphrase" (*Canadian Literature* 70, 1978) condemns this reductive and thematic approach, which obsessively pinpoints recurring themes—such as the "garrison mentality," the predominance of victim figures, and a vision of Canada as a new Eden complete with a fall and redemption—and neglects structural and formal studies. Subsequent critical works, however, for the most part have continued this approach and have tried to map a tradition (John Moss, *Patterns of Isolation,* 1974; Tom Marshall, *Harsh and Lovely Land,* 1979). More illuminating is the regional orientation of Laurence Ricou's *Vertical Man, Horizontal World; Man and Landscape in Canadian Prairie Fiction* (1973), W.H. News's *Articulating West* (1972), and Patrick O'Flaherty's *The Rock Observed: Studies in the Literature of Newfoundland* (1979). A greater critical range can be found in the collection of essays edited by Eli Mandel, *Contexts of Canadian Criticism* (1971), and in the journals *Canadian Literature* and *Essays on Canadian Writing.* (K.M.)

CANADIAN LITERATURE IN FRENCH

The French language in Canada. The valley of the St. Lawrence River, first explored by Jacques Cartier during his second voyage to North America in 1535, was colonized by France during the 17th and 18th centuries. When New France was ceded to Britain in 1763 the Roman Catholic population of more than 60,000 persons spoke a language that was already a blend of several French dialects, although French was then not yet standardized in France itself.

After the British victory over New France, immigration from France virtually ceased, but the number of French-speaking inhabitants continued to increase. About five-sixths of Canada's Francophones live in the province of Quebec and refer to themselves as Québécois. Their literature, called French-Canadian or *littérature québécoise,* records their common history and expresses their unique identity.

The French régime, 1535–1760. During these two centuries not a page of French was printed in New France; there was no printing press in the colony until after the establishment of British rule. The substantial colonial literature written in and about New France was published in France for a European audience. It included accounts of discovery and exploration, official reports and correspondence, travelers' narratives, annals of missions and religious communities, and histories of the colony. These manuscripts and printed texts were supplemented by an abundant oral literature composed of folk songs, folktales, and legends. Considerable scholarly effort has been expended on the recovery and study of the rich heritage of oral traditions and literary works surviving from the French period.

The importance of newspapers

After the British conquest, 1760–1830. After the devastation of the Seven Years' War, intellectual life was for a time inconceivable. During the first 70 years of British rule, journalism was vitally important to the French-speaking majority. The bilingual *Quebec Gazette* (1764) and, later, French-language newspapers like *Le Canadien* (1806) and *La Minerve* (1826) offered the only medium of mass communication, of contact with Europe and the United States, and of political expression at home. The first scattered indications of literature (anecdotes, poems, essays, and sermons) appeared in their pages, as did the verses and songs of two French immigrants, Joseph Quesnel and Joseph Mermet. Quesnel, French Canada's first significant writer, also composed dramatic texts for amateur actors: his comedy *Colas et Colinette* (1788) was revived as a radio play in 1968.

The literature. *Early literature, 1830–60.* Publication of French-Canadian literature in Canada began in the 1830s. The first collection of verse appeared in 1830; the first novel in 1837; and the first comedy and first tragedy soon afterward. This coincidence of dates reflected in part the gradual organization of primary and secondary education and the increasing availability of French books and periodicals even before the resumption of commercial relations with France in 1855. More important was a growing sense of national identity, apparent in the campaigns for responsible government that preceded and followed the ill-fated rebellion of 1837. The principal publication of the time, François-Xavier Garneau's *Histoire du Canada* (1845–48; *History of Canada*), embodied the new spirit, as did the patriotic poetry of Octave Crémazie, a bookseller in Quebec City.

Movement of 1860

The Quebec Movement, 1860–95. Under the Act of Union (1840), the peripatetic parliament of Upper and Lower Canada moved in 1859 to Quebec City, accompanied by its attendant clerks, public servants, and journalists. The capital, being also the seat of the newly founded Laval University (1852), was an ideal setting for French Canada's first literary grouping, known as the Mouvement Littéraire de Québec or the Movement of 1860. Its dozen members shared patriotic, conservative, and strongly Roman Catholic convictions about the survival of French Canada. Their spokesman, Abbé Henri-Raymond Casgrain, promoted a messianic view of the spiritual mission of French Canadians in North America, now that postrevolutionary France had fallen into what he perceived to be godlessness and materialism. Only a few French Romantic writers were admired and imitated. Philippe Aubert de Gaspé's historical romance of the period of British conquest, *Les Anciens Canadiens* (1863; *The Canadians of Old*); Antoine Gérin-Lajoie's colonization novel, *Jean Rivard* (1862–64); and numerous collections of verse by Pamphile Lemay (*Les Vengeances,* 1875) and Louis Fréchette (*La Légende d'un peuple,* 1887; "The Legend of a People") illustrate the nostalgic and didactic preoccupations of the time. More original works were nevertheless attempted: Eudore Evanturel's *Premières poésies* (1878) broke with conventional imagery, and Quebec's first woman novelist, Laure Conan (the pen name of Marie-Louise-Félicité Angers), published a remarkably sophisticated psychological novel, *Angéline de Montbrun,* in 1881–82 (Eng. trans. 1974).

The Montreal School, 1895–1935. By the end of the century Montreal had become the province's commercial metropolis, and the next literary movement was founded there by Jean Charbonneau and Louvigny de Montigny in 1895 with the École Littéraire de Montréal (Montreal Literary School). The society, with some 60 members, weekly meetings, and elected officers, continued to exist, although intermittently, for nearly 40 years. Its members published extensively, mostly in verse; organized four large public sessions in 1898–99; and issued two collective volumes of their writings, in 1900 and 1925. Their literary doctrine was eclectic, although chiefly influenced by the Parnassian and Symbolist movements in France and Belgium.

The Montreal School included the first French-Canadian poet who can be compared with his French contemporaries. Émile Nelligan, indisputably a genius, composed all his poetry during his teens (1896–99) before lapsing into insanity. His intricate sonnets and rondels were published in 1903 by the critic Louis Dantin (the pen name of Eugène Seers).

Aestheticism versus regionalism

During the first decade of the 20th century two main groups emerged from the Montreal School, the aesthetes (*exotistes*) and the regionalists. The aesthetes, among them René Chopin, Marcel Dugas, Paul Morin, and Robert de Roquebrune, had studied in Paris and were fascinated by contemporary French literature and culture. They founded a short-lived artistic magazine, *Le Nigog* ("The Harpoon"), in 1918, but they remained a tiny minority, often denounced as dilettantes. It was the regionalists (Gonzalve Desaulniers, Albert Ferland, Charles Gill, and later Alfred DesRochers, Claude-Henri Grignon, and Blanche Lamontagne-Beauregard) who became the dominant group over the next 30 years. Their preference for local subject matter and language, as expressed in their magazine *Le Terroir* (1909) and encouraged by the critic Abbé Camille Roy, complemented the French-Canadian nationalism then being promulgated by Henri Bourassa and Abbé Lionel Groulx. Paradoxically, the regionalists were proposing rural and agricultural themes when Quebec society was becoming urban and industrial. The French novel *Maria Chapdelaine* (1914) was grudgingly accepted, but Quebec authors like Rodolphe Girard (*Marie Calumet,* 1904) and Albert Laberge (*La Scouine,* 1918; *Bitter Bread*) who portrayed country life too realistically were censured and ostracized. The one poet who anticipated future trends, Jean-Aubert Loranger (*Les Atmosphères,* 1920), was ignored.

As the educated urban population increased between the two world wars, there was a discernible expansion of intellectual life, modestly supported by the provincial Liberal governments in power from 1905 to 1936.

World War II and the postwar period, 1935–60. By the mid-1930s Canada's economic depression, Quebec's socioeconomic development, and European political events were making Quebec's regionalist literature obsolete.

Urban fiction

In fiction Jean-Charles Harvey attacked bourgeois ideology in *Les Demi-Civilisés* (1934; *Sackcloth for Banner*). Three years later Msgr. Félix-Antoine Savard's *Menaud, maître-draveur* (*Boss of the River*) deplored in lyrical language Anglo-American takeovers of Quebec's natural resources, and in 1938 Ringuet (Philippe Panneton) traced the decline of Quebec's rural economy in *Trente arpents* (*Thirty Acres*). After the interruption of the war years (1939–45), French-Canadian fiction became resolutely urban. Gabrielle Roy's portrait of a Montreal working-class district in *Bonheur d'occasion* (1945; *The Tin Flute*) won the Prix Fémina, and Roger Lemelin's *Les Plouffe* (1948; *The Plouffe Family*) spawned a popular television serial.

Not all novelists were attracted to social realism, however. Some, like Robert Charbonneau, André Giroux, or Robert Élie, wrote first-person introspective fiction influenced by the Roman Catholic novelists of France. Others, like Germaine Guèvremont in *Le Survenant* and *Marie-Didace* (1945 and 1947; translated and published together as *The Outlander,* 1950), continued to examine rural society, though with greater detachment. One of the most prolific novelists, Yves Thériault, found new subjects among Quebec's Eskimo and American Indian peoples (*Agaguk,* 1958; *Ashini,* 1960).

In poetry Hector de Saint-Denys Garneau's unrhymed metaphysical poems (*Regards et jeux dans l'espace,* 1937; "Glances and Games in Space") introduced a new era. Four poets subsequently dominated the 1940s and '50s: Garneau, Alain Grandbois, Anne Hébert, and Rina Lasnier. Although each employed distinctive techniques and images, all expressed their sense of solitude, alienation, frustration, or despair. Each, especially Grandbois, influenced younger writers; for the first time poets of Quebec, rather than poets of France, served as models for the next generation—the Hexagone poets.

A literary association and publishing house, L'Hexagone (1953), became the major force in Quebec poetry for the next 15 years. It published dozens of elegantly printed volumes of verse, launched literary magazines such as *Liberté* (1959; "Liberty"), and organized annual conferences of Quebec and international writers. Its leading figures, Gaston Miron, Roland Giguère, Gilles Hénault, Fernand Ouellette, Jean-Guy Pilon, and Michel Van Schendel, were both theoreticians and practicing poets, writing interpretive essays as well as polished poems.

Theatre in Quebec

Simultaneously, Quebec theatre assumed its modern form. A Montreal company, Les Compagnons de Saint-Laurent (1937–52), created a taste for professional performances of contemporary French plays. Two playwrights, Gratien Gélinas and Marcel Dubé, began writing in colloquial language about the problems of living in a society controlled by the Roman Catholic church and by a paternalistic Union Nationale government. Permanent theatres and professional companies sprang up, their personnel often supported by part-time work with Radio-Canada or with the National Film Board of Canada.

By mid-century Quebec intellectuals knew that their traditional society was out of step with the postwar world. In 1948 the painter Paul-Émile Borduas repudiated the past in a revolutionary manifesto, and in 1950 Pierre Elliott Trudeau and others founded *Cité libre* ("Free City"), a journal of social and political criticism. The "quiet revolution" was not far away.

Separatism's influence on French-Canadian literature

The "Quiet Revolution." During the 1960s Quebec society underwent the greatest upheaval of its history. A new Liberal government set about modernizing the province, revamping the educational system, and creating a powerful Ministry of Cultural Affairs. Campaigns for the independence of Quebec were launched by separatist organizations that coalesced in the Parti Québécois (1968), which became the provincial government in 1976. Writers calling themselves *Parti pris* (Decision Taken) founded a magazine (1963–68) and a publishing house to press their demands for a secular, socialist, and independent Quebec. Intellectuals became vocal and assertive, and literary production more than tripled during the decade.

In poetry the territory of Quebec (referred to as *le pays*) was rediscovered in Paul-Marie Lapointe's *Choix de poèmes: arbres* (1960; "Selection of Poems: Trees") and Gatien Lapointe's *Ode au Saint-Laurent* (1963; "Ode to the St. Lawrence"). Nationalism adopted revolutionary language in Paul Chamberland's *Terre Québec* (1964), and personal rebellion triumphed in the avant-garde magazines *La Barre du jour* (founded 1965) and *Les Herbes rouges* (founded 1968). A preoccupation with freedom of expression (*la parole*) revealed itself in titles like Giguère's *L'Âge de la parole* (1965) and Yves Préfontaine's *Pays sans parole* (1967). Perhaps the most influential collection was Miron's *L'Homme rapaillé* (1970; partial Eng. trans., *The Agonized Life*), a poetic record of the search for a Quebec identity. Michèle Lalonde's ironic "Speak White" (1970) had a similar impact.

During the 1970s poetry was less engaged and more experimental: the concerns of American counterculture were adopted in the works of Raoul Duguay and Claude Gauvreau, and the concerns of French feminist writing were treated by Nicole Brossard and France Théoret, often with marked formalist tendencies.

As published poetry became more esoteric, the general public turned to chansonniers, popular singers, such as Gilles Vigneault or Pauline Julien. The "Quebec song," which in many cases had political overtones, became the poetry of the people.

Fiction became the most widely read genre. Works of fiction reflected the upset of the Quiet Revolution in their radical, often sexual, themes and in their unconventional structures, derived in part from the French *nouveau roman* of the previous decade. The Quebec "new novel" began with Jacques Godbout's *L'Aquarium* (1962) and reached its high point in the brilliantly convoluted novels of Hubert Aquin that followed his *Prochain épisode* (1965; Eng. trans. *Prochain Episode;* "Next Episode"). The same year Marie-Claire Blais's *Une Saison dans la vie d'Emmanuel* (1965; *A Season in the Life of Emmanuel*), which won the Prix Médicis, presented a scathing denunciation of Quebec rural life, and Godbout's *Salut, Galarneau!* (1967; *Hail, Galarneau!*), described the Americanization of Quebec. Constantly renewing himself, Gérard Bessette moved from ironic realism in *Le Libraire* (1960; *Not for Every Eye*) through interior discourse in *L'Incubation* (1965; *Incubation*) to symbolic narrative in *Les Anthropoïdes* (1977). The poet Anne Hébert achieved success with her novel *Kamouraska* (1970; Eng. trans., *Kamouraska*) and won the Prix Fémina for *Les Fous de Bassan* (1982; *In the Shadow of the Wind*). Réjean Ducharme in *L'Avalée des avalés* (1966; *The Swallower Swallowed*) and other novels presented the disenchantment of young people in the nuclear age. Other popular novelists of the 1970s were Jacques Ferron, who poked fun at Quebec institutions; the author and publisher Victor-Lévy Beaulieu, with his continuing saga of the Beauchemin family; Roch Carrier, who mocked biculturalism in *La Guerre, yes sir!* (1968); and Antonine Maillet, creator of the immortal Acadian charwoman (*La Sagouine,* 1971; Eng. trans. *La Sagouine;* "The Slattern") and recipient of the Prix Goncourt for *Pélagie-la-charrette* (1979; *Pélagie: The Return to a Homeland*). The subsequent success of Yves Beauchemin's *Le Matou* (1981; "The Tomcat") and similar books suggested a return to favour of "action fiction."

Innovations in Quebec theatre

The contemporary period has seen an impressive growth in Quebec theatre and dramatic writing, with several dozen original plays being performed each year. This "new Quebec theatre" can be dated from the success of Michel Tremblay's *Les Belles-Soeurs* in 1968 (Eng. trans. *Les Belles-Soeurs;* "The Sisters-in-Law"). It is characterized by experimental approaches, including improvisation and collective creation; by proletarian language (*le joual* in plays by Jean-Claude Germain and Jean Barbeau); by parody (Robert Gurik, *Hamlet, prince du Québec,* 1968; "Hamlet, Prince of Quebec"); and by audience participation (Françoise Loranger, *Double jeu,* 1969; "Double Game"). The leading dramatist of the 1970s was Michel Tremblay, who presented various aspects of alienation, often symbolized by homosexuality. The most significant contemporary development has been the emergence of women's theatre, performed by groups such as the Théâtre Expérimental des Femmes and featuring controversial plays like Denise Boucher's *Les Fées ont soif* (1978; "The Fairies Are Thirsty") or Jovette Marchessault's *La Saga des poules mouillées* (1981; "The Saga of Wet Hens").

With the increase in publication subsidies and publishing activity in Quebec after 1960, production of new literary works eventually exceeded 500 titles per year. Hundreds of anthologies, bibliographies, biographies, and critical studies have appeared, making French-Canadian literature one of the most thoroughly documented and exhaustively analyzed literatures of the Americas. Critics like Jean-Éthier Blais, André Brochu, Gilles Marcotte, and Pierre Nepveu, together with literary historians like John Hare, Laurent Mailhot, and Paul Wyczynski, have studied the traditional literary genres, while new forms of paraliterature (science fiction, detective novels, comic strips) are constantly being recognized as legitimate objects of investigation.

(D.M.Ha.)

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(K.M./D.M.Ha.)

Cancer

Cancer in humans is a complex of diseases characterized by uncontrolled multiplication and disorganized growth of the affected cells; it may arise in any of the body's tissues. Cancer cells infiltrate and destroy adjacent tissues, eventually gain access to the circulatory system, are transported to distant parts of the body, and ultimately destroy the host. Concomitant with their capacity for unrestrained growth, cancer cells and the tissues they constitute lose their normal appearance, as viewed through a microscope, and assume aberrant functions.

Not all abnormal growths are malignant; those that are not are referred to as benign tumours. In contrast to malignant growths, benign tumours consist of an orderly growth of cells that often are identical to or very closely resemble their normal counterparts. They are not aggressive and do not invade surrounding tissues, spread to distant sites, or kill the host. Such tumours are usually surrounded by a capsule of fibrous connective tissue.

Cancer was known in antiquity. Malignant tumours have been found in Egyptian and Pre-Columbian mummies, about 5,000 and 2,400 years old, respectively. They are documented in ancient medical writings, such as the Edwin Smith and Ebers papyruses, both written about 3,500 years ago. Cancer is also seen in other species, such as domestic animals, birds, reptiles, and fishes. Although growths that resemble and behave like cancer occur in invertebrates and plants, their precise equivalence to cancer in higher living forms remains to be unequivocally established.

Statistics of cancer incidence and mortality show striking geographic differences, varying significantly from country to country; and within countries differences of incidence and mortality occur between the sexes, various ethnic groups, and various occupations. Epidemiologic analysis of such statistics continues to provide valuable clues and insights into the myriad factors that appear to be involved in the causation of cancer.

This article examines the main causes of cancer, the spread of cancer, the various types that are known today, and the methods of treatment that are being used.

The article is divided into the following sections:

CAUSES

It is now known that cancer can be caused by a variety of factors acting either singly or in concert. These include a wide variety of chemical substances, various types of ionizing radiation, and various classes of viruses. This knowledge has accrued from a composite of epidemiologic studies of cancer in humans and from experimental studies in the laboratory. Although much is known about how cancer is caused, the precise mechanism or group of mechanisms involved continues to elude researchers.

Chemicals. Chemicals numbering in the hundreds are known to induce cancer in laboratory animals, and some of these have also been shown to be carcinogenic for humans. While it is well established that long exposure to certain chemicals gives rise to cancer in humans, it is most difficult to determine accurately what proportion of human cancer is due to such exposure. The difficulty arises from the fact that the length of time between exposure and the appearance of cancer is usually prolonged, lasting some 20–30 years; that exposure is more often than not to a variety of chemicals, so that identifying the carcinogen may be difficult, if not impossible; and that carcinogenesis, *i.e.,* the induction and development of cancer, is a process that involves many factors and many phases.

Experiments on laboratory animals have established that the majority of carcinogenic chemicals are not capable of inducing cancer in their original form. Chemical carcinogens are toxic compounds that are foreign to the host. Once they gain entrance into the body by ingestion, inhalation, or absorption, they are modified by metabolic processes in the host's tissues to forms that are less toxic and water-soluble. The modification is an attempt by the host to detoxify the foreign chemicals and eliminate them by excretion. This capacity to detoxify such substances is an adaptive property of a wide spectrum of living systems, including humans, which affords them a selective advantage to exist in a hostile environment. As in most cases, however, this protective property is not foolproof, and in the process of detoxification certain forms of the carcinogen may arise that are even more toxic and active than the parent compound. In this instance the host renders a less toxic compound into one that causes injury to its cells and may eventually cause them to become cancerous. Highly reactive forms of such chemicals interact with vital macromolecules in the host's cells, causing them to be chemically altered. It is believed that when deoxyribonucleic acid (DNA), the genetic material of cells, is so altered that its expression is accompanied by uncontrolled growth, the cell undergoes a transformation that eventually leads to the development of cancer.

Alteration of DNA

DNA appears to be the most probable macromolecule whose alteration leads to cancer, because once cells become malignant their transformed behaviour is passed on to subsequent cell generations, indicating that it is a heritable change. There is evidence, however, that alterations of nongenetic macromolecules capable of regulating cell growth may also be involved in the causation of cancer.

The alteration of DNA is one phase of carcinogenesis. The second phase is a prolonged one during which the genetically altered cell loses its ability to grow in a regulated fashion. During this phase the altered cell, not yet expressing its malignant potential, apparently is influenced by other factors, including noncarcinogenic chemicals; dietary components, such as fat; or substances produced by the host, such as hormones. Since this phase is prolonged and subject to manipulation, it is a major focus of contemporary cancer research; it is hoped that persons at high risk of developing cancer (presumably with cells genetically altered by a chemical, but not yet cancerous) may be treated to prevent overt cancer from developing. Certain chemical substances, such as antioxidants and vitamin A, are particularly promising in this regard.

Despite the fact that cancer is a prevalent disease in modern society, quantitative experimental evidence in animals suggests that, following the exposure of cells to a carcinogenic chemical, the ultimate development of cancer is a relatively rare event. This is no doubt due to the fact that some genetic mutations lead to the death of the affected cells, so that no cancer can develop. A second, more important, reason is due to the ability of cells to "repair" altered or damaged DNA. In this process the segments of damaged DNA are excised, and identical single strands of the required segments are synthesized; these are then spliced into the defect, reestablishing its continuity. If the synthesis has accurately copied the DNA segment as it was prior to damage, there will be no adverse effects. If, on the other hand, cell replication occurs before the damage is repaired, the altered DNA is copied and the damage is amplified. When such cells undergo division the genetic defect is passed on to their descendants. It is reassuring to know that repair processes appear to be uniformly present in living cells and are very efficient, rapidly repairing damaged DNA with great fidelity.

Occupational chemicals. The association of "substances" in the environment with the development of cancer was suggested in 1775 by Sir Percivall Pott, who reported the frequent occurrence of cancer of the scrotum in chimney sweeps chronically exposed to flue dusts. Pott suggested that this might be due to the prolonged exposure of scrotal skin to the dust. Since then, there have been numerous instances in which the development of cancer has been associated with specific occupations.

Numerous carcinogenic chemical hazards have been identified in a variety of extraction and processing industries. These include polycyclic hydrocarbons present in coal tar and its derivatives, such as pitch, tar oils, and creosote, and in products of the combustion and distillation of coal, oil, shale, lignite, and petroleum. In the past, skin cancer took the lives of many long-time workers in these industries. Fumes inhaled by workers during coke-oven operations and in refineries have been associated with high incidences of lung cancer. Benzene, a product of coal-tar distillation, may affect blood-forming tissues and is suspected of being a carcinogen capable of inducing leukemia. Various metals have been implicated as carcinogens for the lung and several other body sites among copper-ore miners and smelters of nickel and cobalt ores. Beta-naphthylamine, a chemical once widely used in the manufacture of aniline dyes, has been shown to be a carcinogen for the urinary bladder.

Asbestos and lung cancer

Asbestos has been established as a carcinogen for the lung and for the mesothelium (membrane) that lines body cavities. Workers chronically exposed to dust containing asbestos fibres have an incidence of lung cancer 10 times the normal rate. (It is noteworthy that the risk of cancer is increased 90-fold in asbestos workers if they also smoke. This is an excellent example of two agents acting synergistically to induce cancer at a higher incidence and often in a shorter time than either agent does alone.) Although a serious effort has been made to reduce industrial exposure to asbestos, in some parts of the world occupational pollution remains responsible for a debilitating and progressive form of lung disease (asbestosis), as well as cancer, in asbestos workers.

Environmental chemicals and pollution. The environment, which includes the atmosphere, land, seas, lakes, and rivers, reflects the activities of society. Environmental pollutants include the myriad effluents of daily living, industrial as well as naturally occurring. Among the greatest atmospheric pollutants are the gaseous and particulate emissions that range from the massive outputs of industry and motor transport, measuring millions of tons annually, to individual puffs of cigarette smoke.

Smoking and cancer

Cigarette smoke has been shown to contain numerous compounds that are known to cause cancer in experimental animals and that appear to be strongly linked to human cancer, especially cancer of the lung. In addition, tobacco smoke has been implicated in the causation of cancer of the mouth, and to a lesser extent the esophagus, pancreas, biliary system, and urinary bladder. Cigarette smoke also has been shown to contain a number of cocarcinogens, substances that appear to enhance the effect of carcinogens when they are administered concomitantly.

Increasing evidence has accrued to suggest that the release of polychlorinated hydrocarbons and certain insecticides into the environment may pose a carcinogenic hazard. Some of these compounds have enjoyed such widespread use that significant areas of land and bodies of water have been contaminated. This, coupled with the fact that such compounds, once ingested by humans, are stored in body fat and released and metabolized very slowly, has made their removal from the environment extremely important, even though it is a difficult, prolonged, and expensive undertaking.

The effect of environmental pollution can best be illustrated by describing the phenomenon of bioconcentration and its impact on the food chain. Widely used insecticides, which metabolically degrade very slowly and are highly soluble in fat, are a case in point. After being applied on land for agricultural uses, such compounds are washed by rains into streams, rivers, and lakes, where they are ingested by microscopic life forms, which serve as food for fish that are, in turn, the major food source for larger fish and aquatic birds. Since these compounds are soluble in fat, after being ingested by an animal they are stored and concentrated in the animal's body fat. Repeated feeding eventually leads to high concentrations of the compounds in the animal's body, so that its subsequent ingestion by a larger predator, perhaps by humans, presents the predator with a significant level of the compound. This can pose a health hazard so serious that, for example, fishermen on a contaminated lake are warned not to eat the fish they catch. The magnitude of the problem is illustrated by the insecticide DDT, high levels of which persist in the environment and in the bodies of humans, despite the fact that its use was restricted worldwide by the early 1970s.

Food additives and cancer

Food additives are another source of environmental chemicals that has caused concern. Although these have been the object of dispute and have given rise to the "natural food" fad, there is no evidence that food additives cause human cancer. In fact, some food additives, especially those that protect foods from becoming rancid, have actually been shown to prevent chemically induced cancer in experimental animals. An additive that remains a matter of concern, however, is sodium nitrite, which is widely used to preserve processed meats. It has been shown that nitrite can react in the stomach with amines, which arise from the digestion of meat, to form nitrosamines, a group of compounds that are potent carcinogens for certain laboratory animals. These compounds are formed in such minute amounts in the stomach that some researchers doubt that they pose a significant carcinogenic hazard for humans. Similarly, the demonstration that certain compounds formed by the burning of meat are carcinogenic for animals must be placed in proper perspective. Carcinogenesis experiments in animals usually require continuous exposure to high levels of chemicals to obtain statistically valid results in their relatively short life span of a year or two; extrapolation of these results to the effect on human health should be approached most carefully.

Problems of greater importance are the naturally occurring carcinogens, which constitute an important hazard in certain environments. One of these, aflatoxin, is formed by *Aspergillus flavus,* a mold that is widely distributed and is a frequent contaminant of improperly stored nuts, grains, meals, and certain other foods. In certain areas of Africa the high incidence of liver cancer in humans appears to coincide with their ingestion of foods highly contaminated with aflatoxin. Such correlations must be interpreted conservatively, however, since these populations are also often plagued by viral hepatitis B, which also has been linked to liver cancer. In Japan and parts of China similar correlations exist, but again their interpretation must be tempered because of the presence of viral hepatitis B and other factors.

Radiation. The carcinogenic effects of ionizing radiation first became apparent at the turn of the 20th century with reports of cancer of the skin in physicians who pioneered the application of X-rays and radium to medicine. Since then it has been well established that ionizing radiations of all forms, including ultraviolet light, are carcinogenic.

Occupational radiation. Numerous examples of radiation carcinogenesis have been documented in occupational settings. Well known among these were women who developed bone cancer as a consequence of the chronic ingestion of radium salts that resulted from licking brushes to a fine point while painting luminous watch dials. Following ingestion, radium is deposited in the mineral component of bones; and the deposition of radium in the bones is sufficiently high to make them autoradiographic—that is, their image can be reproduced on X-ray film simply by placing the bones in contact with the film. Another notable example is uranium mine workers, who develop lung cancer to a significantly higher degree than the general population. The cause has been traced to the inhalation of the radioactive gas radon, released from trace amounts of radium in uranium ore.

Environmental radiation. The environment contains three major natural sources of radiation: radioactive elements in mineral deposits, ultraviolet light from the Sun, and cosmic rays. The carcinogenic potential of radioactive elements and ultraviolet light has been established, while that of highly energetic cosmic radiation remains to be documented. Clearly, chronic exposure to intense sunlight is a major cause of skin cancer in humans; incidence is high in farmers, sailors, and habitual sunbathers. Since the most effective natural screen for ultraviolet light is the natural skin pigment, melanin, individuals with large amounts of melanin—blacks, for example—are resistant to the carcinogenic effects of ultraviolet light. Fair-complexioned people, on the other hand, are quite susceptible. It is important to point out that the term skin cancer, as it is generally used, includes not only malignant tumours of the nonpigmented cells of the skin but also tumours of the pigmented cells (melanoma). Under the proper conditions ultraviolet light can cause cancer to develop in the very cells that produce the pigment that affords protection from solar radiation.

Sunlight and cancer

The environment also contains dangerous ionizing radiation from artificial sources. These include X-rays used for medical diagnosis and therapy, radioactive chemicals, radioactive elements used in atomic reactors, and radioactive fallout arising from the testing of nuclear devices. Home appliances have been known to emit potentially harmful X-rays under certain circumstances, as was the case with some colour television sets made during the 1950s. Radiation leaks from improperly constructed or operated microwave ovens can also occur, but the health hazards from such exposures have yet to be established.

Certain medical applications of X-rays, such as those used to establish the size and position of a fetus before birth and to control acne, have been largely abandoned because of the increased risk of cancer. Modern technology has greatly reduced the risk of diagnostic X-rays, however. During therapeutic X-irradiation for the treatment of cancer, great care is taken to focus radiation on the tumour and to shield the adjacent normal body tissues from undue exposure. Laboratory workers who use radioactive chemicals are careful to prevent undue exposure by contamination and to dispose of radioactive wastes properly. The cancer risk from radioactive fallout resulting from the testing of atomic devices is virtually impossible to establish with certainty because of the large geographic areas affected. In general, such exposure is not massive, and immediate effects are probably not of high significance. The absorption by growing crops and farm animals of long-lived radioactive isotopes, such as strontium, from fallout may pose a threat to humans, however, because the crops and animals are a source of food.

Viruses. Numerous viruses have been identified that can induce cancer in every major class of vertebrates, including lower vertebrates such as fish. The evidence for the viral causation of cancer in humans long remained strong but only circumstantial; it is now clear, however, that Burkitt's lymphoma, nasopharyngeal cancer, and T-cell leukemia are almost surely caused by viruses.

Oncogenic (cancer-causing) viruses are classified according to the type of nucleic acid they contain. A common feature of oncogenic viruses is that they induce malignant transformation of target cells, which, upon transplantation into suitable animal hosts, exhibit true autonomous growth, local invasion, and metastases to distant sites. Another common feature is physical integration of virus-specific genetic material in the DNA of the host cells. In the case of DNA tumour viruses, the viral genes are integrated directly, whereas with ribonucleic acid (RNA) viruses the RNA is first transcribed into DNA, which is then integrated. Furthermore, RNA tumour viruses are frequently replicated by the cells they transform, while DNA tumour viruses are not. The complexities of tumour viruses are such that progress to establish cause and effect has been quite difficult. The following viruses are strongly suspected of being oncogenic for humans.

Herpes-type viruses. A herpesvirus has been found to be closely associated with two cancers in humans: Burkitt's lymphoma, a malignant tumour of lymphatic tissue first described in children in East and Central Africa; and nasopharyngeal carcinoma, a squamous-cell carcinoma of the posterior part of the nasal cavity, which occurs in high incidence in Chinese originating from South China. Biopsies from both tumours grown in culture gave rise to cell lines in which a DNA virus of the herpes class was identified; it was named the Epstein-Barr (EB) virus. Patients with either type of cancer have high levels of antibody directed against EB virus, indicating that they have encountered the virus and responded to it. Human blood lymphocytes can be transformed in culture by EB virus, and there is experimental evidence that EB virus is also oncogenic for cottontop marmosets and owl monkeys. In addition, EB virus is considered to be the causal agent for infectious mononucleosis, a distressing but self-limited viral infection well known in North America; yet infectious mononucleosis is not followed by either Burkitt's lymphoma or nasopharyngeal carcinoma. It is not known why EB virus causes a self-limited disease in one instance and a variety of cancers in others. Evidence suggests that genetic factors may be important for development of the malignant tumour.

Burkitt's lymphoma

A second herpesvirus that is becoming increasingly suspect as a human oncogen is the herpes simplex virus type 2 (HSV-2), a close relative of the virus that causes the common fever blister. The epidemiology of squamous-cell cancer of the uterine cervix is entirely consistent with the possibility that it may be caused by an agent that is transmitted during sexual intercourse. A woman who develops a herpes infection of the cervix is likely to develop cervical cancer, and women with cervical cancer often have high levels of HSV-2 antibody. Other viruses that may be involved and are known to be transmitted by sexual contact are cytomegalovirus (CMV) and papillomavirus; both are capable of causing cell transformation in culture. The experimental application of radioactive probes of antibodies to specific viral proteins and to components of the viral genome have localized either the integrated genome or the viral antigens of HSV-2, CMV, and papillomavirus in the cells of cervical cancer in women. Although this cannot be taken as unequivocal evidence that the viruses are indeed the causal agents, it does strengthen the hypothesis that cervical cancer may have a viral basis.

Herpes simplex virus type 2 and cancer

The implication of virus as a cause of human cancer has been largely circumstantial, but in one human cancer, thymus-derived (T-cell) leukemia–lymphoma, the evidence is quite impressive. These tumours, which appear in inhabitants of the southernmost island of Japan and in the West Indian population in the Caribbean, have been shown possibly to have an infectious basis. Furthermore, an RNA virus, human T-cell leukemia virus (HTLV), which bears no apparent relation to any of the known animal oncogenic RNA viruses, appears to be present in patients with T-cell leukemia and can be isolated from their tumour tissue with reasonable reproducibility. Finally, infection of normal blood T lymphocytes in culture by HTLV leads to their uncontrolled growth.

Infection with hepatitis B virus (HBV) is endemic in populations that also have a high incidence of liver cancer. This has led some researchers to conclude that the virus is the cause of liver cancer. As noted earlier, however, cancer is a multifactorial disease, and liver cell cancer is an excellent example. HBV infection invariably leads to the augmented growth of liver cells, which renders them exquisitely sensitive to carcinogens. Since populations plagued by a high incidence of liver cancer are also chronically exposed to numerous toxic substances, some of which are carcinogenic for the liver, it is difficult to dissect the interactions of virus and carcinogen with sufficient precision to either indict or exonerate HBV as a liver oncogen. Thus, HBV must be considered only as a possible carcinogen. HBV may well be considered a cocarcinogen, however; it may so increase the sensitivity of liver cells to carcinogens that it decreases the latent period during which the cells undergo malignant transformation and increases the number of tumours that ultimately develop.

Oncogenes. Progress in tumour genetics and molecular biology may lead to a clearer understanding of how cells undergo malignant transformation. As stated earlier, RNA tumour viruses invariably integrate their DNA into the host's DNA. Cells of numerous vertebrate species, including humans, have such integrated DNA in their genes. The conservation of such genes throughout so many diverse species suggests that they probably serve an important role in cells, perhaps in growth regulation. The fact that such genes are almost ubiquitous in their distribution and apparently are simply not expressed in normal cells led to the enunciation of the oncogene hypothesis, which suggests that silent (unexpressed) oncogenes, or genes capable of inducing cancer, can, upon proper stimulation, become expressed and thus cause a previously normal cell to become malignant. Oncogenes have been found in chromosome 8 in Burkitt's lymphoma cells and in a variety of other human tumours. Introduction of such oncogenes in DNA from human tumours into normal cells in culture causes the normal cells to become transformed and behave like cancer cells. Activation of an oncogene in a human tumour has been shown to be accompanied by a minor change in its chemical structure. This suggests that silent oncogenes may be activated by chemicals, radiation, and viruses, all of which are known to alter DNA and to cause cancer. The oncogene may be the common denominator through which such diverse agents act.

Genetic factors. The common cancers of humans are not generally inherited, although some families show an incidence of a particular cancer beyond expectations. Multiple cases of childhood cancer may occur in a family. When leukemia develops in one of a pair of identical twins, the other twin also develops leukemia in about 15 percent of the cases. Among fraternal twins this happens less than 1 percent of the time. Some familial tendencies have also been observed for cancer of the breast, uterus, prostate, stomach, colon–rectum, and lung, although the hereditary effect is not strong. It is possible that familial tendencies reflect common environmental relationships or an environmental factor superimposed on a genetic one.

Leukemia in twins

In studies of familial tendencies for breast cancer in England, Denmark, and the United States, an incidence two to three times that expected in the general population has been noted for close relatives of persons with the disease.

A rare cancer of the eye called retinoblastoma occurs in childhood as a result of the transmission of a single dominant gene (a dominant gene need be inherited from only one parent for the trait to become manifest).

Pheochromocytoma, a rare cancer of the adrenal gland, and carcinoma of the thyroid medulla (the inner portion of the thyroid) sometimes occur in various members of the same family. Either both tumours or only one may occur. Persons with pheochromocytoma also have an increased risk of parathyroid and other rare tumours.

Xeroderma pigmentosum, an inherited, abnormal sensitivity of the skin to ultraviolet radiation, almost always leads to skin cancers. The sensitivity is due to an inherited defect in the ability of cells to repair DNA that has been damaged by ultraviolet light.

Children with Down's syndrome, or mongolism, a disorder associated with an extra chromosome in each cell, have a much greater likelihood of developing leukemia than do other children. Persons with Fanconi's aplastic anemia (in which all types of blood cells are abnormally few and the bone marrow is not fully developed) and some other rare syndromes characterized by greatly increased chromosome breakages also have an increased risk of leukemia.

In chronic myelocytic leukemia (myelocytes are precursors of certain white blood cells), an abnormally short chromosome called the Philadelphia chromosome is a genetic marker in bone marrow cells and white blood cells of some patients. It is not known whether the cancer and chromosomal abnormality are independent events or whether one is the result of the other.

There is increasing evidence that genetic alterations are involved in the causation of cancer, and that these can be either inherited or induced and involve either a minute point in a strand of DNA or a major rearrangement of a chromosome.

Trauma and infection. Repeated trauma (injury) or irritation has been associated with some cancers. Ill-fitting dentures associated with cancer of the mouth and chronic infection of the uterine cervix associated with cervical cancer are examples. In each instance it must be remembered that injury and infection are invariably followed by healing, which involves increased local growth of host cells in the affected tissues, and such cells are quite sensitive to carcinogenic factors. There is no scientific basis to support the notion that trauma alone is a carcinogenic stimulus.

THE SPREAD OF CANCER

The ability of certain cancers to spread from their sites of origin, rendering their treatment and eradication difficult, is the major reason for the sense of hopelessness that is closely associated with them. If it were not for this ability to spread, most cancer could be successfully treated.

The spread of cancer results from either or both of two processes: cancer cells may spread by direct extension from the primary site as a consequence of growth and tumour cell movement invading the surrounding normal tissues, or they may enter the vascular system by invading lymphatics or blood vessels. In the latter process the cancer cells can be transported to sites distant from the primary site, whereas in the process of direct extension the cancer cells may remain close to the site of origin.

Direct extension. The ability of a localized focus of cancer to spread directly is dependent upon its interaction with adjacent tissue. Research has shown that experimental and some human cancers with a high incidence of spread contain enzymes capable of digesting elements of the surrounding normal connective tissues. This allows cancer cells to migrate through the connective tissues, which normally serve as a barrier. Once invasion by local extension has occurred, the opportunity for successful therapy is diminished. There is considerable evidence that tumour cell populations are heterogeneous and differ markedly in their ability to spread. The ability to digest elements of the connective tissue may be limited to a small number of cells in a given tumour cell population.

Metastasis. Metastasis is the process by which cancer cells are spread to sites distant from the primary tumour. The process requires that the cancer cells enter the vascular system, which includes lymphatics, veins, and arteries. Once they are within the system, the cancer cells are transported passively to distant sites. In lymphatics they are trapped in lymph nodes, which may serve as temporary barriers. They may enter the blood vessels by extension from lymph nodes or, more rarely, by passing directly through the connective tissue sheath surrounding lymph nodes. The cancer cells are eventually entrapped in the smallest branches of blood vessels, the capillaries. Cancer cells thus trapped are the "seeds" from which larger colonies of tumour cells grow into secondary sites, called metastases.

It was once thought that metastasis was a random process. Within certain limits, however, the patterns of metastases now appear to be tumour-specific. For example, cancer of the breast and prostate have a predilection for spreading to bone, in contrast to melanoma and kidney cancer, which lodge preferentially in the lung. It appears that the process of lodgement of cancer cells in a distant organ may be related to specific interactions between the cancer cells and those of the organ in which they ultimately settle and resume their growth. Basic research on the nature and mechanism of such interactions may eventually lead to a better understanding of the metastatic process and may result in the development of therapeutic strategies to either limit or totally prevent the spread of cancer to distant organs. (It should be pointed out that the ultimate success of a metastatic focus to establish itself and thrive may depend on other host factors, such as cell-mediated and humoral immunity.) Since the capacity for distant metastasis appears to be a property of a small subpopulation of cancer cells within the primary tumour, prevention or control of metastases may prove to be both feasible and effective in the future.

Effects of metastasis on survival

Spread of cancer and survival. Clinical experience with the relation of the spread of cancer to patient survival is dramatically emphasized by the statistics of cancer of the colon and rectum and of the female breast. In both the colon and the rectum, cancer begins in the cells that line the inner surface of the thin-walled, sausage-like connective tissue structure of the organ. If the cancer cells are localized to and within the wall, 94 percent of patients survive five years after surgical removal of the cancer; when the cancer has spread through the wall, but has not entered the adjacent lymph nodes, five-year survival is 88 percent. On the other hand, once cancer has involved the lymph nodes, the five-year survival decreases to 55 percent. Similarly, in breast cancer in women the 10-year survival rate in the absence of nodal involvement is 72 percent, and when lymph nodes closest to the tumour are involved, 10-year survival is 66 percent. When the tumour has spread to more distant nodes, however, five-year survival decreases to only 31 percent. It is important, therefore, to diagnose cancer and begin treatment before cancer cells begin to spread.

Implications of metastasis for therapy. Just as subpopulations of cancer cells within primary tumours have varying abilities to spread by metastasis, they may also differ in response to treatment by radiation and chemotherapy. This is supported by the observation that metastases may differ significantly in their response to radiation therapy and chemotherapy, and that tumours often appear to become resistant to therapy. In both instances the different response of the cancer cells may be due to a selection process. In the first instance, the differing responses of metastases suggest a coupling of the tumour cell's sensitivity to its capacity to spread widely. In the second, the tumour's decreasing sensitivity to therapy may be due to a selection of more resistant tumour cells by virtue of therapy, in a process analogous to the appearance of bacteria resistant to a particular antibiotic during a course of treatment.

TYPES OF CANCER

Cancers are classified by the type of tissue in which they arise and the cell type that constitutes the tumour. This system of classification recognizes in humans more than 150 types of cancer with different behaviours.

Types of tissue. The organs of the body consist of a variety of tissues, including those that line their inner surfaces (epithelia) and those that render structural support and contain their blood supply (the connective tissues). Tumours of a given type of tissue differ from those of the normal tissues in that and other organs but may closely resemble tumours of similar tissue in other regions of the body.

Carcinoma and sarcoma

Classification of cancers according to the tissue from which they evolve is by two major types: carcinoma and sarcoma. Carcinoma, the larger category, refers to cancers of epithelial tissues, which cover the external body (the skin) and line the inner cavitary structures of organs such as the breast; the respiratory and gastrointestinal tracts; the endocrine glands, such as the pituitary, thyroid, and adrenals; and the genitourinary system, which includes the prostate, testes, ovaries, fallopian tubes, uterus, kidneys, and urinary bladder. Sarcoma refers to cancers of the various elements that constitute the connective tissues, such as fibrous tissues, muscle, blood vessels, bone, and cartilage. Rarely, a cancer is composed of both epithelial and connective tissue simultaneously and is referred to as carcinosarcoma. Leukemias, lymphomas, and other cancers of the blood-forming tissues are classified separately, although strictly speaking they can be regarded as a subset of cancers of the connective tissues. Tumours of nerve tissues, including the brain, and melanoma, a cancer of the pigmented cells of the skin, are also classified separately.

A carcinoma of the intestine has more in common with a carcinoma of the stomach, lung, or breast than with a sarcoma of lymphoid cells of the intestine. It is on the tissue and cells of origin, not the organ of origin, that the peculiar and characteristic properties of a tumour more usually depend.

Squamous-cell and basal-cell carcinomas

Types of cells. Tumours are also classified according to the type of cell from which they are derived. For example, skin cancer includes tumours composed of the major cellular component, squamous cells, giving rise to squamous-cell carcinoma; of the less frequent basal cells found at the base of the skin, whose tumours are referred to as basal-cell carcinomas; and of the still more rare cells that produce melanin pigment, whose malignant tumours are classified as melanomas. Cancers in different organs and individual carcinomas of the same cell type may have different doubling times (*i.e.,* the time it takes for the tumour mass to double in the number of cells or in size). Doubling times observed for lung cancer, for instance, may vary widely, with ranges from eight days to more than 700 days. The doubling time is a reflection not only of the rate of tumour cell growth but also of the loss of cells through spontaneous death.

Paraneoplastic syndromes

In addition to the obvious problems associated with the local mechanical effects of tumours, as well as their spread to distant vital organs such as the lungs, brain, and liver, where their subsequent growth may interfere with function, tumours may also exert other potent biological effects. These tumour-associated phenomena are referred to collectively as paraneoplastic syndromes and may include a wide and bewildering variety of adverse effects, such as loss of appetite, body wasting, fatigue, stupor, coma, excessive thirst, inappropriate flushing of the skin, anemia, spontaneous bleeding, clotting, loss of motor function, dementia, and paralysis. Some of these effects have been linked to the inappropriate production and release of certain hormones by tumours, while other effects remain unexplained. The fact that some effects appear before cancer is diagnosed and disappear as soon as the tumour is removed suggests that many of the unexplained effects are probably also caused by substances released by the tumours. Such effects can be so severe and debilitating that they present a more acute clinical problem than the underlying cancers.

Death in patients with cancer is more often than not due to a superimposed systemic complication, such as an overwhelming infection or uncontrollable hemorrhage. In certain instances death is the result of tumour invasion of contiguous tissues, leading to kidney obstruction or perforation of the gastrointestinal tract.

Sites of origin. Cancers of organs such as the brain, lung, esophagus, stomach, liver, gallbladder, pancreas, kidney, prostate, ovary, and testis grow to form firm masses and hence are referred to as solid tumours. These are contrasted with those tumours involving cells of the blood-forming organs and lymphatic systems, which are normally free and circulating and remain so when they become malignant. The most common sites in which cancers arise in adults include the skin, large intestine, lung, breast, pancreas, prostate, and uterus. Leukemia is the most common tumour of childhood. Some of the more common forms of cancer are discussed below.

Skin. Cancers of the skin are relatively common; they occur in highest incidence on the exposed skin of the head and neck of persons chronically exposed to sunlight. The most common form, about 82 percent of all cases, is that arising from basal cells in the deepest layer of the skin. The initial lesion is a small pimple-like elevation, which enlarges very slowly and after a few months forms a shiny, somewhat translucent lesion that eventually develops a small central ulcer. When the scablike surface is denuded it tends to bleed and then appears to "heal" by forming another shiny covering. Although basal-cell carcinomas grow slowly and only very rarely metastasize, they do invade locally and cause considerable destruction of adjacent tissues, which can result in disfigurement. Either ample surgical excision or radiation therapy is curative.

Squamous-cell carcinoma arises from the platelike flat cells that constitute the major cellular component of skin. The early lesion is less localized and elevated than that of basal-cell cancer. It is red and scaly and may be confused with eczema or infection. Eventually, the lesion becomes larger, elevated, and ulcerated. The behaviour of squamous-cell carcinoma differs somewhat from that of basal-cell cancer, in that the cells are not only capable of local invasion but also may metastasize to regional lymph nodes and, rarely, to more distant sites. Treatment is identical to that for basal-cell cancer.

Lung. Lung cancers arise in the epithelium lining the bronchi (the branching complex of air passages), by which air passes to the lungs, or in the fine air sacs at the periphery. The most common forms arise in bronchial glandular epithelium that has been altered by long exposure to cigarette smoke to form less specialized squamous cells, which eventually evolve into squamous-cell carcinomas. Structurally, unaltered glandular epithelia of bronchi may also undergo malignant transformation to give rise to adenocarcinomas, but these tumours do not appear to be related to cigarette smoking.

Cancers of the lung tend to metastasize widely to lymph nodes in the neck and chest, to the pleura (membrane) lining the chest and lungs, and to the liver, adrenals, and bone. The average survival of persons with untreated lung cancer is about nine months after diagnosis; spread to regional or distant lymph nodes or other sites has already occurred in the majority of cases. In operable cases removal of the tumour may prolong life for a number of months. Five-year survival of patients affected with squamous-cell carcinoma of the lung is about 25 percent and for adenocarcinoma 12 percent.

Risk factors for breast cancer

Breast. Cancer of the breast is the leading cause of death from cancer in women. (The incidence of cancer of the breast is significantly higher in North America and northern Europe than in Asian and African countries.) Women who begin menstruation at an early age are at greater risk than those who begin later, and women who have their ovaries removed surgically before age 35 appear to be at significantly less risk than women who undergo natural menopause. Women who bear their first child before age 18 have one-third the risk of developing breast cancer of those who have their first full-term pregnancy between ages 18 and 30. Treatment is reasonably successful when it is guided by knowledge of the degree of sensitivity of the tumour cells to the estrogenic hormones.

Esophagus. Cancer may arise at any point along the passage through which food is conveyed to the stomach, although the most frequent site is the middle third. The tumours are bulky, fungus-like growths that rapidly close the esophagus; rarely they may spread superficially without causing obstruction. Esophageal cancers are usually squamous-cell carcinomas that invade locally and metastasize rapidly to organs in the chest as well as to more distant sites. Even with the best therapy available, the cure rate is quite low. Chronic smoking and use of alcohol are considered to be major risk factors; diet and certain environmental factors may also be involved.

Stomach. Cancer of the stomach is a major cause of death worldwide. In some countries, such as the United States, the death rate from stomach cancer has decreased dramatically since the 1930s, for reasons that remain unknown. High increased rates of stomach cancer have been documented in coal miners and asbestos workers, in farmers in Japan, in nickel refinery workers in the Soviet Union, and in rubber and timber workers. The cancers almost exclusively affect the glandular epithelium lining the stomach and are adenocarcinomas. The tumours infiltrate and invade the wall of the stomach or form masses that protrude from the surface and ulcerate. These tumours tend to metastasize early and widely. Surgery is the treatment of choice but is curative only when the tumour is diagnosed and removed early. Rarer tumours include lymphosarcomas and tumours of the smooth muscle cells of the wall. These tend to grow more slowly and as a result are more easily cured by surgical intervention.

Liver. Cancer of the liver is a disease whose incidence varies widely; in African countries such as Nigeria and Benin and among Chinese in Singapore, Taiwan, and Hawaii the incidence is high, while in the continental United States and in western Europe primary cancer of the liver is relatively rare. Environmental factors appear to play an important role; the disease is linked to infectious hepatitis B virus, to malnutrition, and to natural chemical carcinogens such as aflatoxin B_1, toxic alkaloids from plants used to brew certain native teas, and nitrosamines. Chronic alcohol abuse, which leads to cirrhosis (scarring) of the liver, is a significant risk factor, especially in the Western world, since it appears to predispose a person to the development of liver cancer. Primary cancers of the liver are carcinomas arising from liver cells, the major cellular component of liver (hepatocellular carcinomas), or, more rarely, from a minor cellular component of liver, bile duct cells (cholangiocarcinomas). Both types of carcinoma tend to spread extensively within the liver; and the hepatocellular carcinomas also grow into the veins of the liver, from which they may spread to more distant sites, such as the lungs. Cancer of the liver may be successfully treated surgically if the tumour has not spread widely within the liver or beyond it.

Risk factors of pancreatic cancer

Pancreas. Pancreatic cancer arising from the epithelium lining the duct system is the most frequent malignant tumour of this organ. In the United States, England, Israel, and several Scandinavian countries the incidence of the disease is high; in other western and eastern European countries and Japan it is intermediate; and in southern Europe and Southeast Asia it is low. Risk factors include cigarette smoking, diabetes mellitus, and perhaps the typical Western diet, consisting of a high intake of meat and fat. Onset of the disease is insidious, so that by the time symptoms and signs such as loss of appetite and weight, jaundice, and painless enlargement of the gallbladder due to obstruction are apparent, the tumour has often already involved contiguous tissues and organs. The cancers are adenocarcinomas that metastasize early and often widely. These characteristics are responsible for the lethality of this disease, which has less than a 2-percent five-year survival rate. Despite considerable clinical research aimed at the development of new diagnostic techniques, successful results by surgical resection have been obtained only in a relatively small percentage of patients who have an early diagnosis. Cancer of the endocrine tissue in the pancreas, which secretes hormones such as insulin, can induce severe disturbances in blood-sugar levels that lead to dizziness, confusion, weakness, and loss of consciousness. These tumours are generally not aggressive and are often cured by surgery. Cancer of the acinar cells, which produce the enzymes necessary for the digestion of food, though rare, has a very poor prognosis, closely resembling that of ductal cancer.

Colon and rectum. Colorectal cancer is very common in the Western world, with an equal incidence in males and females. By contrast, its incidence in Japan, many South American countries, and sub-Saharan African countries is very low. Epidemiologic studies suggest that a major risk factor may be the low fibre content and high meat protein and fat content of the Western diet. It has been observed that Japanese immigrants to the United States within a generation acquire a much higher incidence of colorectal cancer than their counterparts in Japan. Similarly, their children, and the children of other immigrant ethnic groups from countries with low incidence, have incidences for colon and rectal cancer equal to those of other Americans. This is an excellent example of the role of the environment in the causation of cancer. Numerous studies aimed at identifying the early precursor lesions seem to indicate that the majority of cases arise from certain types of preexisting polyps and that these may remain benign for many months before becoming malignant. The tumours are adenocarcinomas, and some apparently also grow very slowly, taking as long as six to eight years to reach a size of about 2.4 inches (six centimetres). The tumours are round and raised and may be ulcerated. Once the tumour has grown through the wall of the bowel, successful treatment becomes more difficult. The tumours metastasize to the liver, lung, and other distant sites. Surgery is the most favoured treatment, although chemotherapy and radiation therapy may also be used when warranted by the extent of the disease.

Kidney and bladder. Adenocarcinoma of the kidney, which arises in the epithelium lining the renal tubules, is the most common primary malignancy affecting this organ. It is more common in men than in women and appears during the fifth and sixth decades of life. Although its causes remain to be determined, a number of retrospective epidemiologic studies implicate an association with cigarette smoking. An important manifestation is blood in the urine (hematuria), which may be painless or accompanied with flank pain. Tumours are large and bulky and may occupy a large portion of the kidney. Renal adenocarcinoma spreads by direct extension to adjacent tissues, and metastasis occurs to lung and bone by way of the lymphatics and the bloodstream. Some tumours may become quite large, however, without any evidence of metastasis. The mean survival rate following surgical removal is about 35 percent. A special type of kidney cancer that occurs in infancy and early childhood is often first discovered as a mass by a parent while bathing a child. These cancers, called nephroblastoma, or Wilms' tumour, arise from abnormal embryonic tissue and involve both connective tissue and epithelial cells. They can spread both by direct invasion and by the lymphatics and the bloodstream. Early diagnosis and combined surgical removal and radiation therapy give favourable results, with a cure rate of about 80 percent.

Wilms' tumour in children

In the United States bladder tumours account for about 6 percent of all cancers, while in Zimbabwe, Egypt, and Iraq bladder cancer constitutes about 40 percent of all cancers. The disease is three times more frequent in men than in women. Naphthylamine and other chemicals used in the production of dyes are known to cause bladder cancer. A metabolite of the amino acid tryptophan is also carcinogenic and may represent one of the means by which bladder cancer appears in people having no contact with the chemical industry. Chronic conditions of the bladder, such as infestation by the parasite *Schistosoma,* found in the Middle East and Africa, and developmental defects that predispose a person to the formation of bladder stones and infection are also considered high-risk factors. Hematuria is the most important symptom associated with this disease. Tumours often begin as benign lesions (papillomas), which become progressively more aggressive, with a high tendency to recur, and finally assume the behaviour of malignancy. The majority are carcinomas of transitional-cell epithelium, which lines the bladder; a much smaller number are squamous-cell cancers, which arise as a consequence of the modulation of transitional epithelium to squamous epithelium by factors and mechanisms that remain obscure. The tumours can be treated successfully with early diagnosis and vigorous treatment. Once they invade the bladder wall and enter adjoining structures, however, the prognosis is poor.

Prostate. Cancer of the prostate is a common disease, but markedly low incidences are seen in Oriental populations, especially in Japan. As with other tumours, immigration from low-incidence to high-incidence countries is followed by an increased incidence in the migrant population. Although no definite etiologic factors have been established, the isolation of HSV-2 from cells of prostatic cancer and the presence and persistence of CMV virus in cultured human prostatic cells have suggested that these viruses might be involved. The fact that prostatic carcinoma is rare before the age of 50, increases in incidence in subsequent years, is not seen in castrates, and regresses following castration implicates aging and the presence of male hormone as significant factors. Cancer arises in the epithelium lining the prostatic acini and small ducts, and, more rarely, in the main ducts. The majority are adenocarcinomas that tend to infiltrate the organ stroma and break through the capsule to spread to the rectum, the base of the bladder, and eventually to more distant sites by dissemination via lymphatics and blood vessels. Metastases to the bones are quite common and, together with involvement of nerves in the pelvis, are often the cause of considerable pain. Treatments include castration, administration of diethylstilbestrol, and, as the extent of involvement may require, surgical removal, radiation therapy, and chemotherapy.

Uterus. Cancers of the uterus are relatively common and represent about 19 percent of all malignant diseases in women in the United States, as contrasted, for example, to Thailand, countries of the Far East such as Japan, and African countries, where they are relatively rare. The two major types of uterine cancer are adenocarcinoma of the lining of the uterus (endometrium) and squamous-cell carcinoma of the womb or cervix. Endometrial carcinoma has its highest incidence late in the sixth decade. The disease is associated with obesity, diabetes, hypertension, and late menopause. Etiology has not been established; estrogens probably play a significant role. Endometrial cancer spreads both superficially and by invasion of the wall of the uterus. Lymphatic spread occurs late. Inappropriate bleeding from the fifth decade of life onward is an important symptom that must be heeded early if treatment of the disease is to be effective. Treatment of choice is radiation therapy followed by surgical removal, and the cure rate is good. Treatment for disease that spreads beyond the uterus includes hormonal therapy with progesterone and, less frequently, chemotherapy.

Major types of uterine cancer

Squamous-cell carcinoma of the uterine cervix is a somewhat more common disease than endometrial cancer and tends to occur at a younger age, beginning as early as the third decade. A viral etiology is strongly implicated and is supported by a higher incidence of the disease in women who have had an early and active sexual history, including multiple pregnancies. Such cancers are infrequent in women who have not been pregnant and are very rare in celibate women. An early symptom of the disease is abnormal bleeding. The cancers arise in the squamous epithelium of the womb and tend to remain confined to the lining for a number of years in the in situ stage before becoming invasive. Women who have an annual examination, including a Papanicolaou (Pap) test, a painless sampling of cells from the cervix, have an improved chance of early diagnosis. Once the tumour becomes invasive, involving both direct extension and lymphatic metastasis, the prognosis becomes grave. Serious complications, such as infection and eventual failure of the kidneys as a consequence of obstruction and fistula formation, occur in late stages.

Ovary. Cancer of the ovary kills many women of the industrialized countries of the Western world. It is relatively uncommon in Far Eastern countries, especially Japan, and in less developed countries. Immigrants to Western countries show higher incidences about 20 years after their immigration, suggesting the possible role of environmental factors; no firm associations have been established, however. Many different types of ovarian cancer have been identified; they arise either in the epithelium or in connective tissue components, very rarely in both. The most common forms are adenocarcinomas arising in the epithelium. Some of these retain their capacity to produce estrogenic hormones and, more rarely, masculinizing hormones. The tumours may be treated successfully through surgery if they are diagnosed early. Extension of the disease beyond the ovary requires more radical treatment, such as radiation therapy and chemotherapy.

Incidence of cancer of the ovary

Lymphoid tissue. Malignant diseases of lymphoid tissue (lymphomas) are the seventh most common cause of cancer death in the United States, and certain forms are common throughout many countries of the world. In African countries and on New Guinea, Burkitt's lymphoma is common, while in the Western countries it is relatively rare. Certain lymphomas in humans have been closely linked with viruses; other established lymphomas are linked to significant chronic exposure to ionizing radiation and to inherited immunologic deficiency or to such deficiency induced to prevent rejection of organ transplants or resulting from treatment for certain diseases. Malignancies of lymphoid tissue may arise in one or more of the organs rich in such tissue, including the lymph nodes, spleen, bone marrow, and thymus. More rarely they can arise in organs containing small amounts of such tissue, including the stomach, intestines, testis, and breast. Lymphomas are aggressive malignancies that tend to spread to distant organs and become systemic in their distribution. As with other malignant tumours, early diagnosis and aggressive treatment by radiation therapy or chemotherapy, or in many instances the two in combination, have resulted in an increasing number of cures.

Blood-forming tissues. Leukemias are a heterogeneous group of malignancies of the blood-forming (hematopoietic) tissues, which include the bone marrow, lymph nodes, and spleen. The acute form of the disease in adults is rapidly fatal, with infiltration of bone marrow and other hematopoietic tissues by the malignant cells, while the acute form in children has yielded to treatment and a number of cures are documented. The disease has a worldwide distribution, and the best established etiologic factor involves chronic exposure to ionizing radiation. Other factors that have been implicated include congenital disorders associated with increased chromosomal fragility and instability, such as Down's syndrome, Bloom's syndrome, and Fanconi's anemia; viruses; and certain chemicals and drugs. Leukemic cells tend to disseminate through the bloodstream, lodging in blood-forming and other tissues and organs. The cells proliferate to the extent that they often crowd out normal elements of the blood, so that patients afflicted with leukemia are rendered anemic by an interference with red blood cell production or from bleeding because of abnormalities of the blood-clotting mechanisms. They are rendered susceptible to infection by a diminution of the various types of white blood cells, which comprise a major defense system of the body. Treatment involves attempts to correct these complications by transfusion of normal blood and blood products, as well as the use of antibiotics to combat infection and of chemotherapeutic agents to destroy the malignant cells. Chronic, more slow-growing, and less aggressive forms of leukemia may continue for years and may respond favourably to radiation therapy, surgical removal of the spleen, and chemotherapy. In terminal cases leukemia cells either show accelerated growth or are transformed to more primitive cells that tend to disseminate widely and cause death within a few months. (D.G.Sc.)

TREATMENT

Successful treatment of cancer requires the complete removal or destruction of all cancerous tissue. If therapy fails to remove all of the cancer cells, the disease recurs. Surgery and radiation are the most effective forms of treatment. Chemotherapy—treatment with drugs and hormones—has proved helpful in some forms of cancer. The choice of therapy is governed by the type, location, size, and extent of invasion and metastases of the cancer at the time of diagnosis, and by the general condition of the patient.

Surgery. For surgery to be curative, it must be performed before the cancer has spread into organs and tissues that cannot be safely removed. Since the late 19th century increasingly radical operations for cancer have become standard. Despite the increasing extent of these procedures, risk has been reduced by improvements in surgical techniques, anesthesiology, and preoperative and postoperative care, especially in the control of infection. Heart–lung pumps, artificial kidneys, and methods of maintaining electrolyte balance and metabolic equilibrium have permitted patients with impaired cardiovascular and kidney functions or poor general metabolism to survive cancer surgery.

Reconstructive techniques

Major advances have been made in the restoration of structures altered by cancer surgery and in the rehabilitation of people who have undergone radical surgery. Patients undergoing certain surgical procedures for cancer of the colon or rectum, for instance, can be equipped with simple devices for the elimination of solid waste. For patients with cancer of the head and neck, the use of grafting methods and of tissue flaps make it possible to apply reconstructive techniques at the time the cancer is removed.

Rehabilitation of the patient also plays an important role. Women who have extensive surgery for breast cancer are given treatment for restoration of muscle tone needed for movement of the arms. Progress has also been made in teaching new mechanisms of speech to people who have undergone surgical removal of the larynx.

In addition to saving lives by eradicating cancer, surgery also may improve the remaining months or years of life for persons whose cancers cannot be eradicated, restoring comfort and a sense of usefulness. When severe pain accompanies cancer, surgery may bring relief by severing the nerve pathways that carry the painful sensations. In addition, surgery is sometimes necessary to treat abscesses resulting from either the tumour or infection and to relieve intestinal obstructions.

Surgery is also valuable as a preventive measure in controlling cancer. It may be used to eliminate precancerous conditions in the mouth, chronic ulcers (ulcerative colitis) that may lead to cancer of the colon, and certain precancerous polyps in the colon and rectum. It may be used to remove burn scars that may lead to cancer, precancerous nodules in the thyroid gland, and certain precancerous pigmented moles.

Radiation therapy. Radiation therapy makes use of ionizing radiations—X-rays, particles (electrons, neutrons, and pi-mesons), and gamma rays—to destroy cells by impairing their capacity to divide. Although some normal cells are also killed during radiation therapy, this occurrence is minimized by careful shielding of the adjacent areas.

Some cancers do not respond to radiation therapy. The differing sensitivities of various malignant tumours to irradiation are due primarily to the variations in the cells of origin of the tumours. In addition, individual cells within a tumour may have a widely different susceptibility to irradiation. In cancers composed of a variety of cells in different stages of differentiation (*e.g.,* epidermoid carcinoma), even a large dose of radiation may not affect the most differentiated cells. Poor circulation within some tumours decreases their oxygen supply, further diminishing their sensitivity to radiation.

Development of instruments that produce energy in the range of millions of electron volts has permitted extensive use of radiation therapy. Such instruments can deliver a greater radiation dose to deep-seated tumours without the serious skin reactions and discomfort often associated with lower energy X-ray beams. Large areas can be irradiated with more precise definition of the margins and alignment of the beam, less side-scatter, and therefore more protection of adjacent vital structures. The greater versatility of such equipment has made possible the development of treatment techniques involving multiple intersecting radiation beams, rotation of the patient or the radiation source, large fields of radiation shaped to the contours of particular organs, and the pinpointing of beams for cancers of the eye and the larynx. Because of their increased usefulness, instruments (linear accelerators, betatrons, and radioactive cobalt-60 teletherapy apparatus) have become standard instruments for radiation therapy of deep-seated cancers.

Chemotherapy. Chemotherapy can cure certain forms of cancer. Cancers frequently cured by drugs include choriocarcinoma, a rare, highly malignant tumour that originates in the placenta; acute leukemia of childhood; and Burkitt's lymphoma. Treatments with combinations of drugs have produced long-term, disease-free remissions in many children with acute leukemia and in persons with advanced stages of Hodgkin's disease. Some of these patients have been in remission longer than five years and may, in fact, be cured.

Multiple carcinomas of the superficial layers of the skin have been eradicated after the application of certain cancer-drug ointments to the skin. Wilms' tumour, for example, is often curable when cancer drugs are used in combination with surgery and radiation. Many other forms of cancer benefit temporarily or partially by chemotherapy. Some cancers are resistant to drugs, however.

Problems with chemotherapy

Most cancer drugs are limited in their usefulness. One problem is that only a certain proportion of cells is dividing at any one time, and most cancer drugs can destroy only that part of the cell population undergoing division. Another problem is that cancer drugs damage normal, as well as cancer, cells and tissues. In addition, some cancer cells eventually become resistant to drugs. To help overcome these difficulties, combinations of chemotherapeutic agents that act on cells in different ways have been used in various treatment programs simultaneously or in sequence. This allows cancer cells at different stages of division to be killed, causes less damage to normal cells, and diminishes the tendency of malignant cells to become resistant to a single drug.

Special measures have been developed to protect persons undergoing cancer chemotherapy from the combined effects of the drugs and the disease. Refinements in the transfusion of blood platelets permit multiple transfusions in leukemia patients with critical platelet deficiencies caused either by leukemia or drug toxicity. During periods when a patient's white blood cells are depleted, administration of antibiotics offers protection from many infections. Relatively germ-free environments, such as specially designed hospital rooms, provide sterile atmospheres to protect against fatal infection. Transfusion of type-matched white blood cells, and, more recently, transplantation of bone marrow cells from carefully matched donors, have also been of value in protecting persons with nonfunctioning bone marrows against infections. (C.G.B./D.G.Sc.)

Experimental treatments. Since the 1970s considerable effort has been expended toward the development, control, and elimination of cancer by immunotherapy. Methods have involved the use of immunotherapeutic agents such as Bacille Calmette-Guérin (BCG), a vaccine against human tuberculosis; killed suspensions of several types of bacteria; chemical products isolated from certain bacteria; and interferon, a family of proteins that inhibit the growth of viruses. All of these substances appear to stimulate the immunologic defense system. It may someday prove possible to cure cancer by manipulating the complex immunologic mechanisms of the host.

Hyperthermia

A second experimental approach involves the use of hyperthermia (high temperature). Research indicates that cancer cells are more sensitive to the killing effects of high temperature than are their normal counterparts. The success of treatment by hyperthermia appears to depend on the development of new instrumentation that will allow the application of highly localized heat to the site of a tumour.

Two developments in radiation therapy that may add significantly to its effectiveness include the use of chemicals that greatly sensitize tumour cells to the damaging effects of ionizing radiation, while not affecting normal cells, and the application of particle-beam radiation therapy, using beams of protons, helium ions, and heavy ions of such isotopes as carbon-12, neon-20, and argon-40. Such beams have sufficiently high energy to penetrate deeply into tissues; they do more damage to malignant cells than do the less energetic forms of ionizing radiation administered at the same dosages. Such developments may allow treatment of deep-seated tumours without significantly damaging overlying normal tissues. (D.G.Sc.)

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Canton

The largest city of South China, Canton lies near the head of the Pearl River Estuary (Chu Chiang K'ou), more than 90 miles (145 kilometres) inland from the South China Sea and the British crown colony of Hong Kong. Although the city is commonly known as Canton to Westerners—a name that derives from that of Kwangtung Province, of which it is the capital—the true Chinese name of the city since the 3rd century AD has been Kuang-chou (Guangzhou in Pinyin). Because of its position at the meeting point of inland rivers and the sea, it has long been one of China's main commercial and trading centres. It has served as a doorway for foreign influence since the 3rd century AD and was the first Chinese port to be regularly visited by European traders. The city is also a historic centre of learning and, as a centre of political activity for the Chinese Nationalist leader Sun Yat-sen, it was the cradle of the Chinese revolution.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

The city site. The city proper of Canton lies within the county (*hsien*) of P'an-yü, although the municipality (*shih*) of Kuang-chou embraces P'an-yü and several additional counties. The main part of the city is situated on the north bank of the Pearl River, which branches off the Hsi Chiang ("West River") and forms the northern border of the immense Pearl River, or Canton, Delta (*san-chiao-chou*) to the south.

Most of Canton lies to the south of Pai-yün Shan (White Cloud Mountain), which rises to 1,253 feet (382 metres) above sea level about four miles from the city centre. At the southern extension of the Pai-yün Shan are the Yüeh-hsiu Shan, on which the earliest known inhabitants of the region lived. Archaeological work has revealed that the site of the city during the Ch'in (221–206 BC) and the Han (206 BC–AD 220) dynasties was slightly north of the modern urban centre. Later, the city expanded southward as river-borne silt and sand were deposited and the Pearl gradually became narrower.

Urban renewal

Old Canton was a crowded city of narrow streets and winding alleys. A vigorous modernization program was carried out in the 1920s and 1930s, during which wide streets were built, modern sewers introduced, arcades constructed for sidewalk shops, and numerous parks created. New dikes built along the Pearl allowed the city to expand southward to its present waterfront. Limited by hills to the north and by numerous waterways to the west, Canton's subsequent expansion has been mainly into the low plains to the east.

Climate. Canton has three seasons. From April through October, the summer season is wet, hot, and humid; south and southwest winds are often accompanied by typhoons, which are seldom destructive. The July mean temperature is 83° F (28° C). Winter lasts from October through early February and is mild and free of snow; the January mean temperature is 56° F (13° C). The third season, from February through mid-April, is a period of transition that is marked by muggy weather. The average annual rainfall is 64 inches (1,625 millimetres), and farmers in the surrounding country enjoy a year-round growing season.

City districts

The city layout. Canton stretches along a waterfront, which runs east and west along the Pearl. Both the Old City (dating to the Ming dynasty and now mostly in the Yüeh-hsiu district) and the districts of Li-wan to the west and Tung-shan to the east are located on the north bank. Since the demolition of the city walls in the 1920s, these sections have become one city. On the south bank are the industrial suburbs of Hai-chu to the south and Hua-ti to the southwest. Another industrial section is located in the northwestern suburb of Hsi-ts'un.

The Yüeh-hsiu district is the commercial centre of Canton as well as the site of provincial and municipal government offices. To the north is Yüeh-hsiu Park, the largest of the city's parks. Within the park are artificial lakes, a five-story red pagoda (built in 1380) that now houses the Canton Municipal Museum, a flower exhibition hall, the Sun Yat-sen Memorial Hall (1931), and sports and recreational facilities. To the west of the park is the Canton Foreign Trade Centre (formerly the Exhibition Centre), from 1956 the site of the Chinese Export Commodities Fair (also called the Canton Trade Fair). The main north–south thoroughfare of the city, Chieh-fang Lu (Liberation Avenue), is intersected in the centre of Canton by the east–west Chung-shan Lu (Sun Yat-sen Avenue). The Peasant Movement Institute, which flourished in the mid-1920s under Mao Zedong's leadership, is located near the city centre. Also in the central part of the district are the Huai-sheng Mosque (built 627), considered to be the oldest mosque in China; the Buddhist Liu-jung Ssu ("Temple of the Six Banyan Trees"), founded in the 5th century, and its nine-story Hua T'a ("Flower Pagoda"); and the Gothic Roman Catholic Cathedral, built in 1860. The district includes the city's major hotels, department stores, and cinemas; traditional Chinese buildings rarely are found except in the hills to the north. Skyscrapers line the banks of the Pearl in the downtown area and ring Hai-chu Square, a huge open space by the river.

Li-wan district, as part of the Old City, has retained traditional-style housing alongside modern buildings. Li-wan Park is located in the southwest part of the district, while to the northeast is Liu-hua Park. To the south, Sha-mien, a tiny island in the Pearl and once an exclusive enclave of the British and French, is noted for its Western-style mansions.

Tung-shan district, east of central Canton, is predominantly residential; many government officials and repatriated Chinese live there. Located in the district are the Martyrs Memorial Park, dedicated to those killed in the uprising in 1927 against the Kuomintang (Nationalists) and a mausoleum for 72 people who lost their lives in an unsuccessful revolt against the Ch'ing dynasty in 1911. Many of the city's institutions of higher learning are located in the district.

South of the Pearl is Hai-chu district. It is characterized by modern residential quarters and industrial centres. The Sun Yat-sen University (founded 1924) is located there.

THE PEOPLE

Canton is one of the most densely inhabited areas in China; most of its residents live in the central district of the city. The people, called Cantonese, speak a dialect of Chinese that is distinctively different from Mandarin. The earliest inhabitants, of Tai or Shan origin, were assimilated by the Chinese (Han) long ago. There are, however, small groups of Manchus and Chinese Muslims (Hui) in the city. A notable demographic feature is the large number of "overseas Chinese" who emigrated to Southeast Asia, the United States, Europe, and other parts of the world. Since the early 1980s, many of them have returned and resettled in and around Canton.

THE ECONOMY

Industry. Since 1950 there has been substantial expansion of such light industries as the manufacture of electronics, textiles, newsprint, processed food, and firecrackers. Heavy industries include the production of machinery, chemicals, iron and steel, and cement, as well as shipbuilding. Smaller plants have also been developed for the manufacture of consumer goods. Canton is celebrated for its many handicraft products, including ivory carvings, jade objects, embroideries, fans, porcelain, and paper umbrellas. Since the late 1970s there has been considerable foreign investment in Canton, and the city has become one of China's principal tourist attractions.

Handicrafts

Trade. Canton is the centre of the trade of Kwangtung and Kwangsi and the adjacent provinces of South China. Products such as sugar, fruits, silk, timber, tea, and herbs are exported, whereas manufactured goods and industrial equipment are transshipped via Canton into the interior. The semiannual Chinese Export Commodities Fair has become an institution of world trade.

Transportation. Buses, bicycles, and automobiles are the principal means of transportation; the city has some of the worst traffic congestion in all of China. The delta is blessed with innumerable canals and creeks; the smaller canals are used by sampans (flat-bottomed boats propelled by oars) and the larger by steamers or motor launches. Canton is a terminus of inland navigation and is also the focal point of coastal and ocean navigation. The port facilities at Canton and at Huang-pu, 12 miles downstream and now a district of the city, have been expanded to accommodate larger vessels.

Canton is served by railroads linking it to Peking, Kowloon, and San-shui (at the confluence of the Hsi Chiang and the Pearl). Kwangtung is the nation's most advanced province in highway development. Major arterial roads link Canton with other large cities in Kwangtung and with Macau. The city's international airport, north of the city, is the largest in southern China.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Canton's municipal government is part of the hierarchical structure of the Chinese government—and the parallel structure of the Chinese Communist Party—that extends from the national organization, through the provincial apparatus, to the municipal and, ultimately, neighbourhood levels. The principal responsibilities of the Canton Municipal People's Congress, the major decision-making body, include issuing administrative orders, collecting taxes, determining the budget, and implementing economic plans. A standing committee selected from its members recommends policy decisions and oversees the operation of municipal government. Executive authority rests with the Canton People's Government, the officers of which are elected by the congress; it consists of a mayor, vice mayors, and numerous bureaus in charge of public security, the judicial system, and other civil, economic, social, and cultural affairs.

The city is divided into districts (*ch'u*), each of which has a district mayoralty. Under the district, there are police substations and street mayoralties. Neighbourhood associations have various functions, including mediating disputes, conducting literacy campaigns, supervising sanitation and welfare, and promoting family planning.

Health. Health conditions have improved dramatically during the 20th century. Epidemics have been eliminated through the control of disease-carrying pests. The city has many fine hospitals, of which one of the best known is the Kwangtung Provincial People's Hospital.

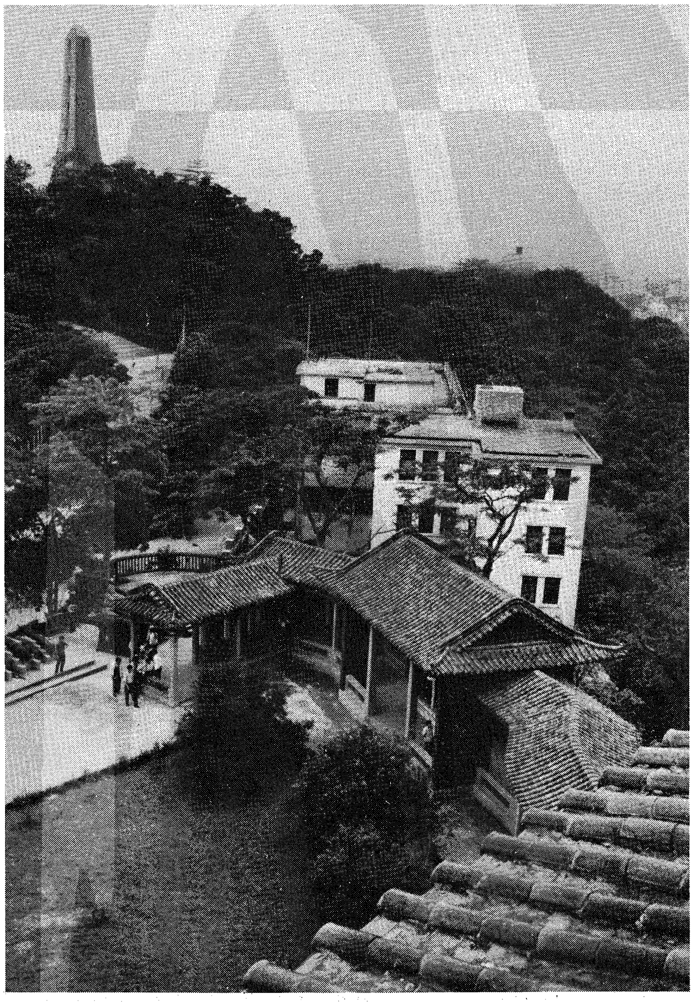
Education. Canton is one of China's most progressive cities regarding education. In addition to a large number of kindergartens, primary schools, and middle schools, it has many institutions of higher learning, including Sun Yat-sen University, Chinan University, Sun Yat-sen Medical College, Canton College of Traditional Chinese Medicine, Canton Institute of Mechanical Engineering, and South China Institute of Agriculture.

CULTURAL LIFE

Canton is a vibrant city. Large tracts of parkland, many of them created since 1949, help give the city its characteristically lush appearance and provide the citizenry with a wide variety of recreational facilities. The colourful flower show held during the annual Spring Festival is a major highlight. Although the Cantonese are increasingly aware of the modern life-style of Hong Kong, they also prize their historical and cultural roots. Throngs of people are usually found visiting the city's many museums and monuments of historical importance, and productions of Cantonese opera and music are well attended. The Sun Yat-sen Library of Kwangtung Province, with one of the largest holdings in China, has an extensive collection of vernacular-language works. The vitality of the Cantonese is exemplified by their passion for food. Cantonese cuisine is world-renowned, and the city's hundreds of restaurants offer a seemingly endless variety of dishes.

Cuisine

Percy L. Julian—Globe Photos/Stockphotos, Inc.



The Yüeh-hsiu Park in Canton. The Sun Yat-sen Obelisk is in the left background.

History

THE EARLY PERIOD

The earliest known inhabitants of the Canton area were the Pai Yüeh, a Tai, or Shan, people. At the beginning of the Western Chou dynasty (*c.* 1111–771 BC), their chief built a walled town, known as Nan-wu Ch'eng, in the northern section of the present-day city. In 887 BC the town was taken by the mid-Yangtze kingdom of Ch'u and was known as Wu-yang Ch'eng ("City of Five Goats").

Under the Ch'in dynasty Canton was made the capital of Nan-hai prefecture. Upon the fall of the Ch'in, Gen. Chao T'o (died 137 BC) established an autonomous state known as Nan Yüeh, which was annexed in 111 BC by the Han dynasty. For the next 300 years Chinese assimilation of the Yüeh people proceeded, and integration of the region into the empire took firm root.

During the four centuries from the Three Kingdoms to the founding of the T'ang dynasty in AD 618, when North China was overrun by "barbarian" invaders, Canton remained a part of the Chinese regimes based in Nanking. During this period the city grew in wealth and population; Buddhist temples were erected, and a flourishing community was maintained by Arab and Hindu traders. Peace and prosperity were further augmented under the T'ang (618–907). An auxiliary wall and settlement were built on Yü Hill, but the city suffered much destruction during the civil strife at the end of the dynasty.

Under the Sung dynasty (960–1279) the increase in Canton's population and the growth of foreign trade made it necessary to enlarge the city. A second auxiliary wall and settlement were constructed on P'an Hill in the late 11th century. With the settlements on the twin hills, the city took on the name P'an-yü (hence the name of the county in which Canton is now located). Under the Southern Sung (1127–1279) Chinese seamen and traders sailed to Southeast Asia, thus opening the way for Chinese emigration abroad in subsequent ages. In the late 13th century and throughout the 14th, many Chinese families from North China moved into the Kwangtung region in the wake of the Mongol conquest. A booming economy resulted as the Yüan rulers (1206–1368) encouraged maritime trade and kept Chinese-Mongol race relations under control.

Period of expansion

Under the Ming dynasty (1368–1644) the city underwent considerable rebuilding and expansion. In 1380 the P'an and Yü hills were razed, and the old town and the two auxiliary districts were combined into one large walled city. In 1535 an outer wall was added to incorporate the new commercial districts on the north bank of the Pearl. Meanwhile, the pattern of foreign trade changed as the supremacy of the Arabs ended with the coming of the Europeans. The Portuguese sent their first embassy to Canton in the early 1500s, followed by the Dutch and the British in the 17th century.

Canton came under the rule of the Ch'ing (Manchu) dynasty from 1644 to 1911/12. Recognizing the importance of the city, the government made it the capital of the Viceroyalty of Kwangtung and Kwangsi. The British East India Company established a "factory" (foreign traders' residences and business offices) in Canton in 1685, and annual trading operations began in 1699. Throughout the 18th century French, Dutch, American, and other nationals also established trade relations with the city; the "13 factories" were located on the waterfront.

Trade moved without undue difficulty until friction began to mount in the 1820s. The foreigners found trade restrictions (through licensed Chinese merchants known as Cohong) too irksome, while the Chinese authorities refused to open normal diplomatic relations. The Chinese seized and destroyed large quantities of illegal opium brought in by the British in 1839, and in retaliation the British attacked Chinese positions in the Canton Estuary. The first Opium War (1839–42) ended in humiliating defeat for China, and the city saved itself from destruction only by paying a $6,000,000 ransom.

The Treaty of Nanking (1842) with the United Kingdom opened Canton as a treaty port. In 1844 the French and the Americans obtained similar treaties. Antiforeign sentiment, however, ran high, and the city refused to open its gates until 1857. The coolie trade and the use of foreign flags to protect pirates caused several crises. The second Opium (or "Arrow") War broke out between China and Britain and France in 1856. Canton was occupied by Anglo-French forces until 1861, and Sha-mien was made an Anglo-French concession in 1859.

Amid the woes of foreign imperialism, Canton was deeply shaken by the great antidynastic outbreak of the Taiping Rebellion (1850–64), the leader of which, Hung Hsiu-ch'üan, was born in the northern Canton suburb of Hua-hsien. Many followers of Hung formed secret societies that kept his revolutionary ideals alive even after the failure of the rebellion. For the next 50 years anti-Manchu agitation formed one of the twin forces that gripped Canton; the other was the rise of nationalism.

THE MODERN CITY

Canton came under the spell of its most illustrious son, Sun Yat-sen, from 1885 to 1925. Sun made the city the testing ground for his campaign to overthrow the Manchu dynasty and to establish a Chinese republic. The Canton Uprising of 1911 paved the way for the success of the revolution before the end of the year. Canton became the base of operations for action against the warlords between 1916 and 1925 and served as the headquarters of Sun's party, the Kuomintang. Besides completing his Three Principles of the People, Sun reorganized the Kuomintang in 1924 to reactivate the Nationalist revolution. All manner of people flocked to Canton—the right- and left-wing members of the Kuomintang, the members of the newly formed Chinese Communist Party, and Soviet advisers. Chiang Kai-shek, Mao Zedong, and Zhou Enlai began their careers in Canton under Sun's tutelage.

Headquarters of the Kuomintang

Chiang gained power when he crushed an uprising by the Canton Merchants Volunteer Corps and defeated the disloyal local warlords (1924). With Sun's death in 1925, however, Canton was embroiled in the power struggle between the Communists and the Nationalists. In 1927 a Communist-led coup attempted to set up a workers' government in the city, only to be crushed by Chiang. From 1928 to 1937 Canton was officially under the control of the Nationalist government, but it was actually controlled by independent leaders, who criticized Chiang's dictatorship and threatened secession from Nanking. In 1937, when war against the Japanese broke out, Canton became a prime target of Japanese air raids. Canton fell in 1938 and remained under Japanese occupation until 1945. Recovery did not begin until the Communist government took control in 1949. The modernization begun in the 1920s continued, although there were periods of hardship during the Great Leap Forward (1958–60) and Cultural Revolution (1966–76). New housing and government offices have been built, heavy industry has been introduced, and the city has developed into one of China's centres of foreign trade.

Post-1949 period

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(P.-c.K./Z.G.)

Cape Town

The city of Cape Town (Afrikaans: Kaapstad) is the legislative capital of the Republic of South Africa and the capital of Cape Province. The city lies at the northern end of the Cape Peninsula some 30 miles (50 kilometres), at its southernmost boundary, north of the Cape of Good Hope. Because it was the site of the first European settlement in South Africa, Cape Town is known as the country's "mother city."

Cape Town has a beautiful setting: parts of the city and its suburbs wind about the steep slopes of Table Mountain (3,563 feet [1,086 metres] high) and neighbouring peaks and rim the shores of Table Bay; other parts lie on the flats below the slopes or stretch southward across the flats to False Bay. The city covers an area of 116 square miles (300 square kilometres).

This article is divided into the following sections:

Physical and human geography

THE CHARACTER OF THE CITY

The city of Cape Town had its origin in 1652 when the Dutch East India Company established a refreshment station for its ships on the shores of Table Bay. The location was magnificent, on well-watered, fertile soil, beneath the precipitous walls of Table Mountain. The indigenous inhabitants provided cattle but not labour, and the company imported slaves, mainly from East Africa, Madagascar, and the Bay of Bengal area. The slaves brought with them something of their culture and, especially the Muslims from the East Indies, their religion. Miscegenation of brown, black, and white took place, but strong racial and ethnic characteristics remained. In 1781 the French established a garrison to help the Dutch defend the city against British attack, and the French presence influenced local architecture and culture. British occupation in the 19th century brought new parliamentary and judicial concepts and freedom for the slaves. Cape Town was the gateway to Europe's penetration of the South African interior, and close ties with continental Europe were maintained.

Today Cape Town is a modern city with high-rise office buildings and pedestrian malls. Although it is a major political and economic centre, its reputation still rests on its beautiful situation between mountain and sea, its cosmopolitan population, and the liberal outlook of many of its citizens.

THE LANDSCAPE

The city site. The first settlement of Cape Town was situated between Table Mountain and Table Bay. It was bounded on the northwest by the ridges known as Lion's Head and Lion's Rump (later called Signal Hill), on the north by Table Bay, on the south by Devil's Peak, and on the east by marshlands and the sandy Cape Flats beyond. The nearest tillable land was on the lower eastern slopes of Devil's Peak and Table Mountain and, farther to the southeast, at Rondebosch, Newlands, and Wynberg. From the fortress that protected the settlement, a track led south past these lands to False Bay on the eastern side of the Cape Peninsula and on beyond Muizenberg and Kalk Bay to Simon's Bay, where the East Indiaman trade ships could find shelter from northwesterly winter gales. The constraints of mountain, sea, and sand shaped the direction of Cape Town's growth, and the pattern was followed in subsequent road and rail construction. A railway line reached Wynberg in 1864 and Muizenberg in 1883, and another line ran eastward from Cape Town across the Flats to the interior.

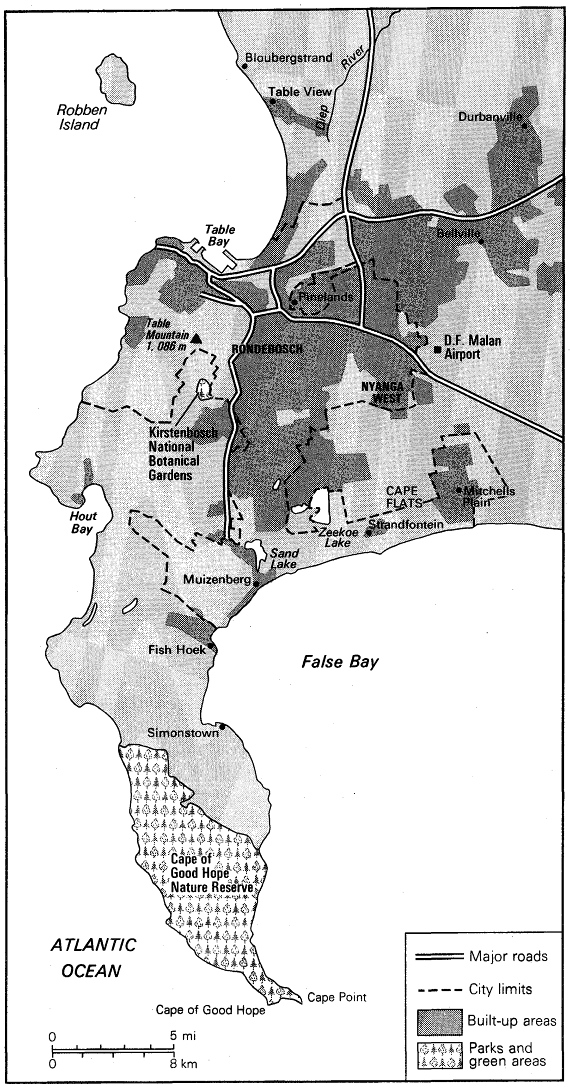
Climate. The climate of Cape Town is Mediterranean in type; it is locally modified by the mass of Table Mountain and by the cold Benguela Current of the South Atlantic. The average high temperature is 70° F (21° C), in January and February, and the average low is 55° F (13° C), in July, but temperatures are cooler on the mountain slopes and on the coast. Freezing occurs infrequently. On the average, rain falls on 69 days of the year; about half of the 26 inches (660 millimetres) of annual rainfall occurs between June and August, the southern winter. The amount of rainfall varies with proximity to the mountain, those areas close to the slopes receiving as much as twice the precipitation of areas farther away. The winds, generally strong, come from the northwest in winter and vary from between southeast and southwest in summer. Southerly winds produce a cloud covering of condensed moisture over Table Mountain known as the "tablecloth." These winds are collectively referred to as the "Cape doctor" because they keep air pollution at a low level.

The "Cape doctor"

The city layout. The narrow streets of the early settlement ran parallel to the shore. The road that led from the shore inland to the Company's produce garden became the main thoroughfare. Originally called the Heerengracht, after the canal in Amsterdam of that name, it was renamed Adderley Street in 1850. Other main roads paralleled it as the town grew. In Strand Street, on what once was the shore of Table Bay, stands the Castle of Good Hope, built by the Dutch East India Company between 1666 and 1679. Near the Castle are the Botanic Gardens, which are bisected by Government Avenue and overlooked by government buildings. A parliament building, constructed for the use of the colonial government and first occupied in 1885, became the seat of the Parliament of the Union of South Africa in 1910 and of the Republic of South Africa in 1961. Additions to the building were opened in 1987.

The opening of the Alfred Dock in 1870 led to renewed development along the shore. The breakwater was lengthened and piers were built in 1890–95, and the newly sheltered area was named Victoria Basin. Dredging for the Duncan Dock, built between 1938 and 1945 to accommodate larger vessels, and for the Ben Schoeman Dock in 1977, resulted in the reclamation of 480 acres (194 hectares) along the shore, referred to as the Foreshore. Adderley Street was extended to the new harbour, and the extension was named the Heerengracht.

The building of a modern system of freeways has greatly facilitated communication between the city and its suburbs. There has been a considerable increase in population and commercial activity in the southern suburbs, especially in Claremont and Wynberg. Particularly favoured



Cape Town and surrounding area.

residential areas are on the slopes east of Table Mountain and on the slopes of Signal Hill, overlooking Bantry Bay and the beaches of Clifton.

Architecture

Much of Cape Town's early architecture reflected prototypes from the Netherlands that were modified for the region. Characteristics included flat roofs, projecting porches, and distinctive gables. Louis-Michel Thibault, a French architect who arrived in 1783, designed much that was then fashionable, including the Old Supreme Court building, now the South African Cultural History Museum. Notable buildings of the period include the Old Town House and the Lutheran Church. Of the original Groote Kerk (Great Church), opened in 1704, only the steeple and vestry remain; the rest was rebuilt in the 1830s. Today, buildings in the central business district and on the Foreshore follow modern international trends in architecture.

THE PEOPLE

Cape Town's population is subject to South Africa's racial classification system, which recognizes the white group, the black group, the Coloured group (legally defined as any person who is not a member of the white group or the black group), and the Asian group (Indians and Chinese). More than half of the residents of the city and metropolitan area are Coloured, about one-fourth are white, about one-fifth are black, and the remainder are of Asian origin. In the metropolitan area Afrikaans is the first language of almost one-half of the Coloureds and whites. Almost one-quarter speak English as a first language, and another quarter are equally at home in both languages. The blacks are predominantly Xhosa-speaking. The majority of the residents are members of Protestant churches, but there are also sizable communities of Roman Catholics and Muslims.

South Africa's Group Areas Act of 1966 consolidated earlier acts aimed at enforcing the policy of racial segregation known as apartheid, and it provided for the reservation of certain areas for residence and occupation by specific racial groups within the population. The act brought about many changes in Cape Town's residential areas. For example, a mixed but predominantly Coloured neighbourhood known as District Six, south of the Castle, was cleared by bulldozers. Special legislation permitted Coloureds who were living in Cape Town's Malay Quarter to remain, but other Coloured and Asian families were forced to move to designated areas, mostly east of the Cape Town–Muizenberg railway line and onto the Cape Flats. According to figures submitted to Parliament, by the end of 1980 some 29,300 Coloured and 1,500 Indian families, but only 195 white families, had been resettled on the Cape Peninsula. Because housing in the prescribed areas was inadequate, in 1975 the city undertook construction with government funding of a model township of 40,000 houses for Coloured families at Mitchells Plain, southeast of the city. Indian families were installed at Rylands and Pelikan Park. In 1990 the government did an about-face over District Six, opening it to residence by all sections of the Cape Town community.

Government legislation controls the presence and employment of blacks outside their designated black states (homelands). Nearly all blacks in Cape Town are confined to Guguletu and Nyanga West within the city limits and to neighbouring Nyanga and Langa. With the abolition of influx controls in the 1980s, a great movement of blacks into Cape Town and other urban areas from impoverished black states began, and camps of squatters were soon overcrowded. The government established a township for blacks at Khayelitsha, east of Mitchells Plain, and squatter camps were then demolished in 1986 in an attempt to direct the blacks there.

THE ECONOMY

Industry and commerce. Cape Town was South Africa's economic base until the discovery and exploitation of minerals in the interior; today it is one of the nation's most important industrial centres and a major seaport. About 90 percent of the fish eaten in South Africa is distributed through Cape Town, and Table Bay is one of the world's largest fruit-exporting harbours. A petroleum refinery and chemical, fertilizer, cement, and automobile-assembly factories are situated in the metropolitan area. In the city the basic industries are connected with ship repair and maintenance, food processing, and wine making and with the manufacture of clothing, plastics, and leather goods. Tourism is of growing importance.

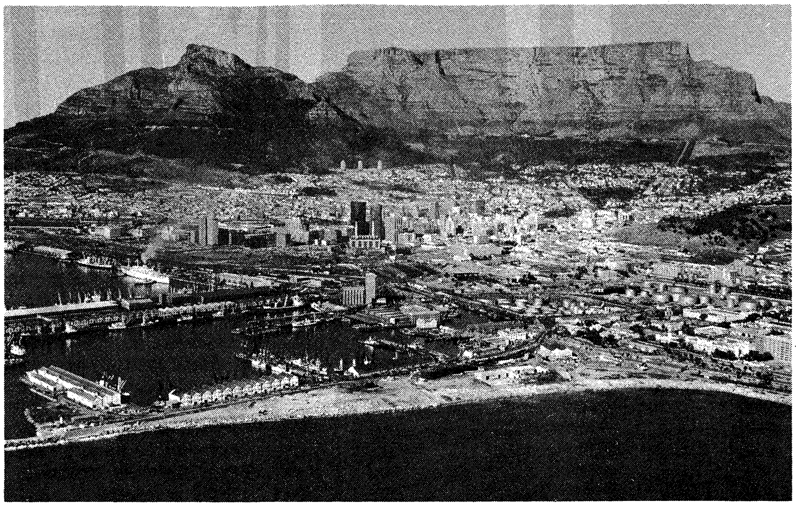
Major industries

Cape Town is well served by department stores and supermarkets. A number of nationwide commercial concerns have their head offices in Cape Town, including oil and insurance companies.

Transportation. The port of Cape Town handles some 5,000,000 tons of cargo annually. The port does not admit ships of more than 40-foot draft at low tide, but its repair facilities and dry dock are important to interoceanic traffic. The Ben Schoeman Dock accommodates container traffic.

The D.F. Malan Airport has regular flights to Europe and South America. The bulk of its flights, however, are domestic. Cape Town is the terminus of a railway network that extends northward to Zimbabwe and beyond.

Two main radial freeways lead southward to False Bay. Supplementing these are two national routes and a beltway circling the central business district. An elevated freeway extends across the city's shore area. Most whites and the more prosperous Coloureds and blacks own motor vehicles. Public transport is particularly important for those forced by law to live far from their place of employment. Spoornet, which operates suburban trains, and a private bus company serve the Cape Town area. There also are minibuses that provide cheap, fast transportation.



Cape Town on Table Bay. Devil's Peak is on the left, Table Mountain on the right.

Shostal Associates

ADMINISTRATION AND SOCIAL CONDITIONS

Government. The city council consists of two councillors from each of 17 wards, elected by adults owning or occupying property. The mayor has largely ceremonial duties, and an executive committee of council members is directly responsible for the administration of the city. Some council decisions are subject to appeal to the administrator of the Cape Province, who assumed the powers of the provincial council upon its dissolution in 1987. At that time local government fell under the Regional Services Council of the Western Cape, on which the Cape Town city council was obliged to serve, although it did so under protest as the Regional Services Council was constituted along apartheid lines.

At one time there were no racial bars in Cape Town, and both whites and nonwhites could vote and hold office. In 1972 (when there were six nonwhite councillors) national legislation removed nonwhites from the electoral rolls against protests from the citizens. The municipality has continued to oppose apartheid legislation and in 1985 formally reiterated its belief that all people, regardless of race, colour, or creed, have the democratic right to participate fully in the affairs of the city and its council. A peaceful protest march of some 40,000 people in September 1989 helped to create the climate for similar marches in other parts of South Africa and for the subsequent release from prison of black nationalist leader Nelson Mandela and for moves away from the system of apartheid.

Services. Most of Cape Town's electricity is produced at the national Electricity Supply Commission's nuclear power station at Koeberg, north of the city. Cape Town has its own coal-fired power station and two gas turbines to assist in emergencies and at peak periods. A hydroelectric facility at Steenbras also generates power when needed. The city's water, which once came from dams on Table Mountain and at Steenbras, now also comes from Riviersonderend, Voëlvlei, and Wemmershoek. The city provides ambulance, fire, and other services to nearby municipalities and local authorities, including black squatter camps, when requested and permitted.

Sources of water

Health. The city's Health Department offers comprehensive health services and runs a system of polyclinics and specialized clinics. The major medical problem in Cape Town, as in every urban centre in South Africa, is pulmonary tuberculosis, a disease that is spread easily where nutrition and hygiene are poor and housing is overcrowded. Public hospitals are the responsibility of the broader provincial administration. Groote Schuur Hospital, where the world's first heart transplant took place, is one of South Africa's largest hospitals. There are other smaller provincial hospitals and an increasing number of private hospitals in the city.

Education. The provincial administration is responsible for the schooling of whites; schooling for blacks is provided by a central government department. Education for Coloureds and Indians is the responsibility of the national House of Representatives (the Coloured house of the parliament) and the House of Delegates (the Indian house of the parliament). Private schools are permitted to accept students of all races, but the fees are high. In 1990 provision was made for the opening of some provincial schools on certain conditions to all races.

The most renowned institution of higher learning is Diocesan College (founded in 1849), located in Rondebosch. The University of Cape Town, also in Rondebosch, developed from South African College (founded in 1829) and formally came into being in 1918. The university has always demanded the right to admit students of all races, conditional only on the basis of academic merit, and an increasing number of nonwhites are being accepted. Lack of space and funding, however, has caused the university to restrict its growth. The University of the Western Cape, originally built to serve the Coloured community, is located in the nearby municipality of Bellville. The facilities of the Cape College for Advanced Technical Education are being centralized in new buildings in District Six; the first complex was opened in 1987. Many residents of the city, especially nonwhites, receive academic degrees through correspondence courses offered by the University of South Africa.

CULTURAL LIFE

The South African Cultural History Museum controls several satellite museums, including Groot Constantia (the manor house built by Governor Simon van der Stel in about 1685), the 18th-century Koopmans de Wet House, the Bo-Kaap Museum (a reconstruction of a Cape Muslim house), the Bertram House (Georgian), and, in the old harbour area, the South African Maritime Museum. There are several collections of art. The William Fehr Collection in the Castle of Good Hope displays paintings and prints related to Cape history, as well as Cape antique furniture, silverware, and glassware. At Rust en Vreugd, an 18th-century house near the Castle, another part of the Fehr Collection displays watercolours and prints of historical interest. Paintings by European masters are on exhibit at the Old Town House, and modern and contemporary paintings are on view at the South African National Gallery. The South African Museum, with extensive additions opened in 1987, is devoted mainly to natural history.

The Fehr Collection of art

The Cape Town Symphony Orchestra was for some time the only municipal orchestra in South Africa. The Cape Performing Arts Board, based in the Nico Malan opera-house-and-theatre complex, has a permanent orchestra, soloists, and chorus. It is known for its opera and ballet productions. The Baxter Theatre, opened in 1977 on the campus of the University of Cape Town, contains a theatre, a concert hall, and a studio theatre, and stages as many as 1,000 performances a year.

Sports and recreation

The cricket and rugby grounds at Newlands are the traditional homes of these sports in South Africa. Soccer matches are held in the Hartleyvale and Green Point stadia, which are also used for cycling and other sports. National tennis competitions are held in Rondebosch. The Good Hope Centre, designed by the 20th-century Italian architect Pier Luigi Nervi, is used for exhibitions, indoor tennis, boxing, and wrestling. Squash courts are located at the University of Cape Town's indoor sports centre. A race track and four golf courses lie within the city boundaries. Zeekoevlei and Sandvlei, two lakes in the southeast near False Bay, are popular for water sports, especially sailing and windsurfing (board sailing). Cape Town's harbour is an important base for yachting; many international yachting events are hosted there. The city has numerous public swimming pools, and its beaches, open to all races, are popular. Table Mountain and the mountains above Muizenberg provide hiking and climbing for nature lovers and mountaineers. The National Botanic Gardens at Kirstenbosch flank a boundary of the city.

History

The first European to anchor at Table Bay and climb Table Mountain was the Portuguese navigator António de Saldanha. He encountered a few hundred indigenous inhabitants, a Khoikhoin people whose economy was based on herding, hunting, and gathering. After Saldanha's visit, European ships continued to put in at Table Bay to take on fresh water, meat, and other provisions. Survivors of the Dutch vessel *Haerlem,* wrecked in Table Bay in 1647, brought back such glowing reports of the region that the directors of the Dutch East India Company ordered that a station to supply ships rounding the Cape be established there. On April 6, 1652, the company's representative, Jan van Riebeeck, stepped ashore to select sites for a fort and a vegetable garden. In 1657 the company began to release men from its employ so that they could become free burghers and farmers, and in 1658 the company began to import slaves. Inland from Table Mountain, a second company farm was established at Newlands, and vines were planted on the slopes of Wynberg ("Wine Mountain").

Van Riebeeck and his senior officials constituted a council of policy and court of justice. Free burghers were invited to join the court when matters concerning burghers were at issue, and burgher-councillors eventually took responsibility for services such as fire protection, road maintenance, and the preservation of order. The colony began to spread beyond the Cape Peninsula, and the council of policy came to rely increasingly on the burgher-councillors for fact-finding and for advice on town affairs.

The importation of slaves, the introduction of political exiles from the Dutch East Indies, and miscegenation with indigenous Khoikhoin (whom the Dutch called Hottentots) increased the population, but at the beginning of the 18th century the town, known as De Kaap ("The Cape"), still consisted of only 200 houses. Its growth was accelerated by rising international tensions and growing appreciation of the strategic importance of the Cape. During the Seven Years' War (1756–63), in which the major European powers were involved, many French and British ships called at the port, which from 1773 onward was referred to by British visitors as "Cape Town." During the U.S. War of Independence, which exacerbated tensions between rival European powers, a British fleet sought in 1781 to occupy the Cape, which directors of the English East India Company described as "the Gibraltar of India." A French fleet, however, reached the Cape first and established a garrison there to help the Dutch defend it. The French presence brought prosperity and gaiety to Cape Town and initiated a surge of building.

Garrisoned by the French

Suburban development, largely inland to the east and down the Cape Peninsula to the south, followed the radial roads and the railway line, construction of which began in 1859. The Alfred Dock, opened in 1870, encouraged shipping. An influx of people followed the discovery inland of diamonds in 1870 and gold in 1886. These developments, as well as the South African War (1899–1902) between Britain and the combined forces of the Boer Republics, brought about a modest industrialization in Cape Town.

The years after World War II were marked by increased urbanization and by enormous growth in both industry and population. The harbour and industrial sites were extended again, and modern buildings rose in the central business district. Handsome residential areas spread along the lower mountain slopes, and modern freeways were constructed.

The degree of local control over Cape Town's affairs has varied considerably throughout its history. Burgher representation on the council increased to six, and from 1785 onward a committee of three burghers and three officials formally advised the council on urban matters. This committee acquired official status in 1793. By the time a British force occupied the Cape in 1795, the committee, known as the burgher-senate, had assumed responsibility for the town, which then had 1,000 houses and a population of some 14,000. In 1803 the colony reverted to the Netherlands. In the following year the town received its coat of arms, embodying an anchor, symbolizing "good hope," on a field of gold, and three golden rings from van Riebeeck's personal coat of arms.

Britain reoccupied the Cape in 1806, and its title was confirmed in 1814. In 1828 the burgher-senate was abolished, and two residents were appointed to the governor's council of advisers. Slaves were freed in 1834, but they were required to serve four years of indenture. In 1840, when the town's population had reached some 20,000, the municipality of Cape Town came into existence; it incorporated the suburbs of Green Point and Sea Point to the north and west of Signal Hill. The municipality was administered by a board of commissioners and ward masters elected in 12 districts by citizens owning or occupying premises valued at £10 or more per annum, and it was financed by property taxes. In 1867 Cape Town obtained full municipal government, with three councillors from each of six districts; the chairman of the council became mayor.

From 1881 onward a number of separate municipalities came into being, among them Woodstock, Rondebosch, Claremont, Newlands, Wynberg, and Kalk Bay. The improved roads, the introduction of an electric tramway, and a common concern for water supplies and sewers prompted proposals for amalgamation of the urban sprawl. It was not until 1913, however, three years after the formation of the Union of South Africa, that the town councils of Cape Town and several of the adjacent towns combined to form the City of Greater Cape Town.

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(E.A.)

Caracas

Caracas, capital of the Republic of Venezuela, is one of the principal cities of South America. It is by far Venezuela's largest urban agglomeration and the nation's primary centre of industry, commerce, education, and culture. Founded in 1567 as Santiago de León de Caracas, the city grew slowly until the 1940s, after which it expanded by monumental proportions and its influence came to extend to even the remotest parts of the country. In addition to its role as the national capital, Caracas serves as the seat of government for the Federal District, comprising an area of 745 square miles (1,930 square kilometres). The city itself, however, spreads across the Federal District boundary, a considerable part of its area lying in the state of Miranda, which borders the district on the east and the south.

The physical setting of Caracas is impressive. Steep mountains of the Venezuelan coastal range rise abruptly from the Caribbean Sea, affording little level land for urban development. About 7 miles (11 kilometres) southward, beyond this mountainous wall, the city spreads out in a valley. Its elevation at the centre of Caracas is 3,025 feet (922 metres), and the valley extends about 15 miles in an east–west direction. Almost all of its inhabitable area is occupied by the city, and this once green and peaceful valley is now crisscrossed by streets and highways.

This article is divided into the following sections:

Physical and human geography

THE LANDSCAPE

The physical setting

The city site. The Andes Mountains extend eastward from Colombia across all of northern Venezuela to the island of Trinidad. The site of Caracas is a high rift valley within this system; hence, mountains enclose the city at all cardinal points. The Cordillera del Litoral, which rises above 8,600 feet at the Pico Oriental and above 7,000 feet at El Avila, towers high above the city and limits its expansion northward. The Serranía del Interior, to the south, is lower but nevertheless rugged. The principal river draining the valley is the Guaire, once relatively large but now reduced in volume and highly polluted. Its principal affluent in the metropolitan area is the Río El Valle, and others include the Quebrada Baruta and Quebrada Anauco.

Climate. Climatic conditions and natural vegetation in the coastal range form an extremely complex pattern. Temperatures generally decrease with increasing elevation, and rainfall and humidity are high on slopes that face eastward toward the prevailing winds. The decisive climatic factor in the valley is altitude. Temperatures range from 45° to 91° F (7° to 33° C), the annual mean being about 70° F (21° C). In contrast, the average temperature at La Guaira, at sea level, just 10 miles away, is 81° F (27° C). The seasonal range of temperature in Caracas between the average of the coldest month (January) and that of the warmest month (May) is only 5° F, considerably less than the normal diurnal range. Annual rainfall totals about 32 inches (813 millimetres). There is a distinct rainy season from May through November, which accounts for more than 80 percent of the total annual precipitation and features warm, humid weather. In contrast is the dry season, from December through April, characterized by clear skies and pleasantly cool temperatures.

The higher mountain slopes, especially to the north, often have a thick cloud cover. There a fairly abundant vegetation, including forest remnants, remains green throughout the year. Other patches of luxuriant vegetation are preserved in the major parks of the city and in its various gardens, including the Jardín Botánico. Where barren slopes prevail, erosion is a serious problem. Bird life, once exceeding 100 species in the valley, has been reduced dramatically.

The old city centre

The city layout. The nucleus of Caracas is near the western margin of the valley adjacent to the hill of Parque El Calvario, which offers a panoramic view of the city. There, centred around the Plaza Bolívar, are many historic buildings reflecting the colonial past, including the Catedral de Caracas, the Capitolio Nacional, the Concejo Municipal (Municipal Council) building, the birthplace of Simón Bolívar, and Palacio de Miraflores, which serves as the official residence of the president of the republic. Only a short distance away is the Panteon Nacional, with the tomb of Simón Bolívar and those of other national heroes. The twin towers of the Centro Simón Bolívar are also located nearby. Once the tallest buildings in the country, these 30-story structures house various ministries of the national government.

Increasingly, the civic centre of Caracas is moving eastward toward the Parque Central and Plaza Venezuela. This district focusses around a monument to Christopher Columbus, but more conspicuous is a cluster of skyscrapers, including another set of twin towers that are among the tallest buildings in all of Latin America. Begun in 1971, these structures of 62 stories each soar to a height of 656 feet above street level. Planned to house the main offices of the national government, each building is equipped for helicopter landings on its roof and affords a spectacular view of the entire valley. Nearby are a number of other important city landmarks, including the Jardín Botánico, several museums, Parque Los Caobos, and the Universidad Central de Venezuela.

Still farther eastward the principal landmarks of the city are the Caracas Country Club, the Parque Nacional del Este, and the Francisco de Miranda Airport. The city has also expanded southward over the hillsides and along valleys tributary to the Río Guaire.

THE PEOPLE

Immigration and the *barrios*

The people of Caracas, like those of Venezuela as a whole, are primarily mestizo. To this mixture of white and native Indian ancestry is added a lesser strain of black African origin. There are few people of Asian background, but since World War II there has been a relatively heavy influx of Europeans. During the 1950s, particularly, large numbers of Spaniards, Portuguese, and Italians settled in Caracas. By 1956, Caracas and the Federal District were the home of 42 percent of Venezuela's 438,000 foreign residents. These immigrants and their descendants have tended to concentrate in certain *barrios,* or neighbourhoods, of the city. Internal migration has added significantly to the population growth rate of the capital city, as has a high index of natural fertility. The primary source of migrants is the western Andean region, particularly the states of Táchira, Mérida, and Trujillo, although Caracas likewise serves as a magnet attracting people from all other parts of the republic.

Social class divisions are quite sharp in Caracas, as elsewhere in South America, and are reflected in the occupancy of residential areas. Members of the upper class first gained wealth and status in Venezuela through landownership and more recently through industry, commerce, and urban real estate. Their homes are luxurious and generally occupy the eastern margins of the valley and the Caribbean coast. The middle class has developed largely in relation to the nation's petroleum wealth and recent European immigration. Middle-class housing occupies the central core of the city, plus certain residential suburbs. The lower social class, composed of labourers, servants, and the unemployed, occupy the hillside shantytowns overlooking the city from the west and south.

Roman Catholicism has been the predominant religion of Caracas since the earliest settlement, but other religions are practiced. Spanish is the universally spoken language, although English is also commonly used in hotels, restaurants, and other tourist-related facilities.

THE ECONOMY

Industry. Caracas is in many ways the centre of everything Venezuelan. This was particularly true in relation to industry and commerce until the 1950s, when the national government began to promote industrial decentralization. It remains the leading manufacturing centre, with emphasis on textiles and clothing; processed foods and beverages; tobacco products; wood, paper, and printing; clay and stone products; rubber and leather goods; glass, chemicals, and pharmaceuticals; and metalware and plastics. More than 20 percent of Venezuela's manufacturing establishments are located within the greater metropolitan area of Caracas, as is 30 percent of the nation's industrial labour force. The relative importance of manufacturing has decreased rapidly, however, as new industries have been developed elsewhere in the republic and those that pollute the environment are required to move from the city to less populated areas.

Centre of commerce and industry

Commerce and finance. Most important in the economic life of Caracas is the tertiary, or service, sector. This includes wholesale and retail trade, transportation and public utilities, education and health care, and especially government. The city serves as headquarters for most of Venezuela's commercial and industrial firms, its banks and insurance companies, and government agencies both foreign and domestic. Nearly all nations of the world maintain embassies or consulates within the city. Tourism is of growing significance to Venezuela and is likewise centred in Caracas, where hotels and tourist agencies abound. In Caracas, as in any city of dynamic growth, the construction trades are yet another source of livelihood. Chronic unemployment nevertheless remains and is only partially relieved by activities such as street vending, peddling, and shoe shining.

Transportation. Caracas is linked directly with other major urban centres of the world by air and ocean transport. Maiquetía Airport, located 10 miles by road from Caracas on the coast, provides international connections as well as domestic flights to all parts of the republic. Two smaller airports, La Carlota and Francisco de Miranda, also serve the city. La Guaira and, to a lesser extent, Puerto Cabello are the natural seaports for Caracas and handle most of its import and export trade. A modern superhighway, including two long tunnels through the mountains, leads from Caracas to Maiquetía, La Guaira, and various beaches and resort areas along the coast. It was opened in 1953 to replace a tortuous 21-mile zigzag route.

Increasing problems of vehicular congestion

Within the city large numbers of buses, private automobiles, and taxis provide transportation, but during rush hours they clog the streets in what has become an increasing problem of traffic congestion. The public transportation system, although deficient, improved greatly with the inauguration, in 1983, of a modern subway system. Traffic also is facilitated by a system of *autopistas,* multilane divided highways extending east–west through the valley and connecting the city with interior locations. Most urgently needed is improved communication with the Caribbean coast, either by adding lanes to the existing *autopista* to La Guaira or by building an additional route farther eastward, a quite costly alternative. Railroads, which once provided Caracas with vital links to the sea and to interior valleys, have been abandoned in favour of highway transportation.

ADMINISTRATION AND SOCIAL CONDITIONS

Government. Caracas is the seat of the national government, which includes executive, legislative, and judicial branches. Among its residents are more than 40 percent of all government employees within the nation. The city itself is governed through the Libertador Department of the Federal District, but the urban area has expanded across many other minor civil divisions of the district and the state of Miranda. The governor of the Federal District, like those of the states and territories, is appointed by the president of the republic and may also be removed by him. A primary responsibility of the Federal District, and of each state, is the preservation of public order within its boundaries. It maintains its own police force and determines which law enforcement agencies will function under municipal control. The Federal District police headquarters is based in Caracas and reports directly to the governor. Also influential in local affairs, since the early days of colonial settlement, is the Concejo Municipal of Caracas. Planning for the city is conducted and coordinated through a Comisión Metropolitana de Urbanismo. Caracas has been the seat of a bishopric of the Roman Catholic Church since 1535 and of an archbishopric since 1804.

Services and health. For almost three centuries of settlement in the valley, local water supplies were considered abundant. Not until 1940 did water shortages become critical, but since that date a continued search for new sources has been required to meet the city's growing demand. La Mariposa and Agua Fría dams were completed in 1949, and in 1956 an aqueduct from the Río Tuy to La Mariposa reservoir brought water for the first time from a source outside the valley. Water supply remains one of the most critical problems to be resolved as the city continues to grow. Also of major concern are waste disposal, contamination, and pollution. Raw sewage is discharged directly into the Río Guaire, and air pollution results from industry and severe traffic congestion within the relatively small, enclosed valley.

Critical concerns of water supply

Caracas has had electric power since 1897 and is now included within a national electrification grid. Natural gas is supplied by pipelines from Guárico and Anzoátegui states in the eastern Llanos region of Venezuela.

In the last half of the 20th century, both government and private investments in Venezuela have focussed on Caracas, with the result that this city has prospered beyond all others. In 1936 the population of Caracas was more than twice that of Maracaibo, the nation's second largest city; by 1971 its population was more than three times greater. Immigrants have come in massive numbers from all other parts of Venezuela, as well as from most European nations. A high birth rate contributes an additional dimension to population growth and to a severe housing problem.

Increasingly, Caracas has become a city of high-rise apartments that provide clean, modern, and attractive housing. Yet, sprawled across the hillsides are countless thousands of *ranchitos,* homes constructed of corrugated metal, cardboard, or other scrap materials. The construction of new housing is impressive but remains totally inadequate to meet the continually increasing demand. Since the 1960s, large blocks of uniform dwellings have gradually replaced old homes in the Spanish colonial style, of which few now remain.

Health facilities in Caracas, like those of most large urban centres, have been challenged by the rapidly growing population, pollution, and the problems of treating a large poor population. Within the metropolitan area there are numerous clinics, sanatoriums, hospitals, and rest homes that provide health care for the capital's inhabitants.

Education. Considerable emphasis has been placed on education in Caracas. There are five sectors of education, both public and private: primary, secondary, normal, arts and trades, and university. The city also has numerous kindergartens and day-care centres for those under six years of age. The leading institution of higher education is the Universidad Central de Venezuela (also known as the Universidad de Caracas), founded in 1725. Its buildings form a modern enclave, called Ciudad Universitaria (University City), in the centre of the valley. Other universities are Simón Bolívar, which is public and oriented toward science and technology, and the private universities Andrés Bello Catholic, Metropolitan, and Santa María. Among other noteworthy educational centres are an institute of hygiene, a polytechnic institute, normal schools for girls and for boys, and a military academy.

CULTURAL LIFE

Like many world capitals, Caracas is the main focus of the country's cultural life. Among the numerous types of cultural activities are theatre, exhibitions of painting and sculpture, concerts, dance, and film. The principal cultural institutions include the Ateneo de Caracas, an ultramodern theatre inaugurated in 1983; the academy of the language; and the academies of history, medicine, and political and social sciences. Also located in Caracas are the national library, numerous other libraries, natural and fine arts museums, art galleries, municipal and national theatres, and an institute of cultural and fine arts.

Centre of Venezuelan communication

Caracas is amply served by radio, television, and the press. Venezuela's first newspaper, the *Gaceta de Caracas,* was published in 1808 and was dedicated to the cause of national independence. The major newspapers of today are still based in Caracas, including *Ultimas Noticias, El Nacional, El Mundo,* and *El Universal.* These papers are distributed throughout the republic by air transport, hence printing and publishing elsewhere in Venezuela are much less developed. There is a wide variety of radio broadcasting from many stations, and the national radio broadcasting organization is headquartered in Caracas; television, first broadcast in 1953, is popular at all sociocultural levels.

Caracas, with its mild climate and excellent facilities, affords its citizens ample opportunity to express their interest in sports and recreation. Baseball and soccer, the leading national sports, are concentrated in Caracas. The city also has a 30,000-seat horse-racing facility, widely renowned for its beauty, and a bullring that is the site of 10 to 12 events per year. There are many golf courses, swimming pools, and tennis courts, plus mountains and beaches along the Caribbean coast. The city fairly abounds with public gardens and plazas and scores of people are attracted to its zoo, museums, historic buildings, and boulevards. The most popular parks are the huge Parque Nacional el Avila (a forest preserve on the Cerro El Avila), Parque Los Caobos, the Jardín Botánico, and the Parque del Este, in La Floresta.

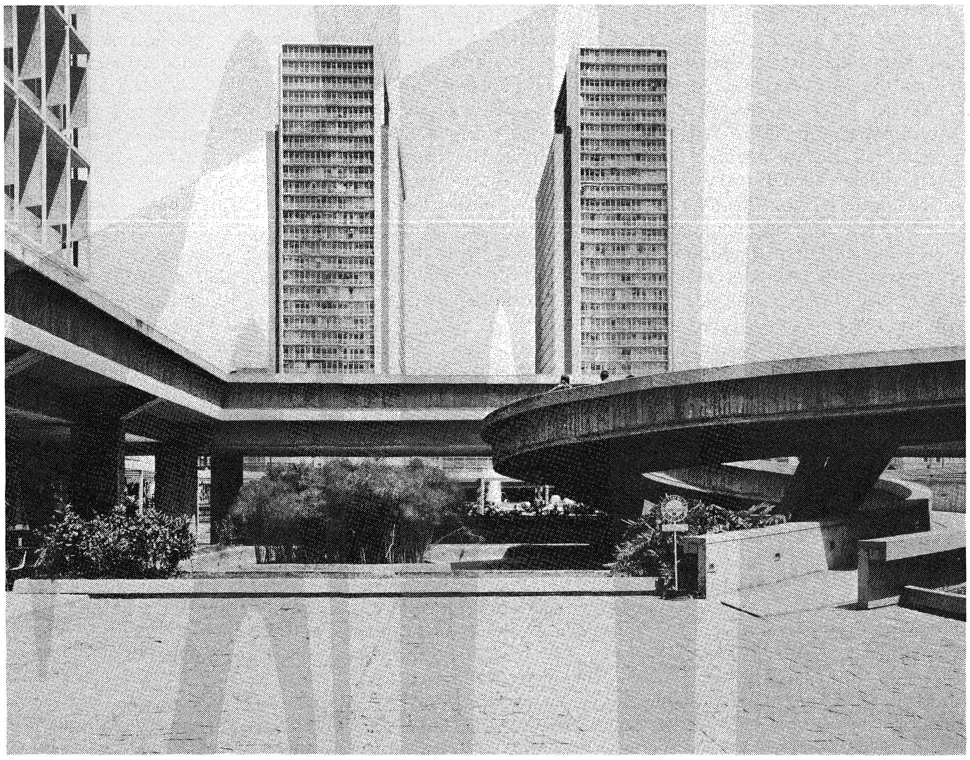
History

Early settlement

The settlement of Caracas occurred more than 40 years after that of Cumaná (1523), in the east, and Coro (1527), to the west. A ranch was established in the valley in 1557 by Francisco Fajardo, the son of a Spanish captain and an Indian chief's daughter, and in 1561 Juan Rodríguez Suárez founded a town on the site of the ranch; but the town was soon destroyed by Indian attacks. The conquest and resettlement of the region began in 1566, and Diego de Losada is credited with the actual founding of the city in 1567. He named it Santiago de León de Caracas in honour of the apostle James, who is the patron saint of Spain; of Don Pedro Ponce de León, who was the provincial governor; and of the Caracas Indians, who inhabited the region.

In 1578 the city was a quadrilateral formed of 24 squares centred on a plaza. The streets were straight and cobbled, and rivulets of water from the hills ran along their margins. Three houses were made of brick, each of two floors, with straw roofs; most houses had walls of *bajareque* (cane and earth), packed earth, or adobe and roofs of straw or tile. During this era, the average house of the upper class was large, with tree-filled patios and arcades and with separate slave quarters and stable. Initial growth was along the road to the north and then to the south. Later, development advanced to the west and east.

The *cabildo,* or town council, of Caracas increasingly exercised authority over the entire central portion of Venezuela, and an elite group emerged to manage the



Eric Carle—Shostal Associates

The older Centro Simón Bolívar in downtown Caracas.

Emergence as the nation's first city

political and economic affairs of the city. In 1577 Caracas acquired its status as a provincial capital. The city was sacked by English buccaneers in 1595, and its buildings were almost totally destroyed by two earthquakes in 1755 and 1812. During the 17th and 18th centuries, however, Caracas gradually emerged as the predominant urban centre of Venezuela. The Indian resistance to settlement was quickly overcome, in part due to an epidemic of smallpox that decimated the native population. The valley was relatively free from pirate attacks, although centrally located near the north coast of the country. The valley itself and adjacent valleys produced some gold and then became important for exportation of cacao, the source of chocolate. In addition, Caracas offered a healthful, comfortable climate and an abundance of fresh water.

Simón Bolívar, the greatest of South American heroes, was born in Caracas on July 24, 1783, into a family long prominent in Venezuelan affairs. From his tutors, he gained the zeal to achieve independence for Spanish America. With a small military force recruited in Colombia, he invaded Venezuela and captured Caracas in August 1813, at which time the city conferred on him the title El Libertador (The Liberator). A year later the patriotic forces were in full retreat, but when Venezuela was finally established as an independent nation, in 1830, Caracas was its capital.

Late 19th-century transformation

In 1870, the urban transformation of Caracas began under Pres. Antonio Guzmán Blanco. When not actually residing in Paris, and ruling Venezuela through a puppet president, Guzmán Blanco was busily trying to make Caracas a replica of the French capital. The French architectural taste of this period replaced the Spanish colonial style. Avenues were built, and some streets were widened to encourage the city's growth. The Caracas and La Guaira Railway was inaugurated, with pomp and ceremony, in 1883. This railroad climbed the steep coastal range and connected the nation's capital with its Caribbean seaport. In the early 20th century, the first urban real estate development, designed for the upper class, was built in suburban El Paraíso. A short time later, a development for the middle class was begun in Catia, to the west.

Beginning in 1936, after the death of Gen. Juan Vicente Gómez, who had been dictator for 27 years, the country experienced a new era of prosperity based on petroleum. Caracas grew in population, and construction was extended as far as geographically possible. In the downtown area of the city, El Silencio district was transformed into a huge middle-income residential-commercial building complex. This development, programmed in 1939 together with a Regulating Plan for Caracas, was the first work of such magnitude in the country; it marked the beginning of the city's modern architectural trend.

Modern Caracas

From 1951 to 1957 Venezuela came under the rule of another military dictator, Marcos Pérez Jiménez, who devoted primary attention to the modernization of Caracas. Slums were cleared and replaced by colourful high-rise apartments, and superhighways were built to connect Caracas with the coast and with the interior. The change in the appearance of Caracas was abrupt and overwhelming. The old city began to disappear. Numerous suburbs sprang up overnight on the hillsides, and commercial and government buildings were erected at a feverish pace in the valley. As early as 1956, a plan was formulated for the decentralization of Caracas and the development of secondary urban-industrial centres. Caracas has continued to grow and develop despite its geographic limitations, and it has come to completely dominate the urban structure of the Venezuelan nation.

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(J.R.Me./C.W.M.)

Caricature, Cartoon, and Comic Strip

Caricature, cartoon (in the satirical sense), and comic strips (including comic books) are related forms. Historically, they arose in the order given, but they all have flourished together. All use the same artistic mediums of drawing and printmaking; caricature and the element of satire are usually present in each. This article covers their history and development.

The article is divided into the following sections:

DEFINITION OF TERMS

Caricature. Caricature is the distorted presentation of a person, type, or action. Commonly, a salient feature or characteristic of the subject is seized upon and exaggerated, or features of animals, birds, or vegetables are substituted for parts of the human being, or analogy is made to animal actions. Generally, one thinks of caricature as being a line drawing and meant for publication for the amusement of people to whom the original is known; the personal trait is usually present.

Origin of the word caricature

The word caricature derives from the Italian verb *caricare* ("to load," "to surcharge" as with exaggerated detail) and seems to have been used first by Mosini in *Diverse Figure* (1646). The 17th-century sculptor-architect Gian Lorenzo Bernini, who was a skilled caricaturist, seems to have introduced the word *caricatura* into France when he went there in 1665. There is perhaps, in the choice of the verb *caricare* as a source for the noun, some influence from the idea of *carattere* (Italian: "character") or even from *cara* (Spanish: "face"). At any rate, the face is the point of departure for most caricatures. It is conceivable that underlying the series of overlapping profiles with varieties of extraordinary noses and chins and brows which Leonardo da Vinci and Albrecht Dürer drew independently about 1500 was an observation not only of contemporary human types but of the fact that the heads of rulers on coins and medals, when worn with age, often became ridiculous. A latter-day case is the penny showing Queen Victoria, whose coiffure began to look like an elephant's head when the coin was well worn down.

Caricature, after its spread as idea and practice from Italy and France to Great Britain in the 18th century, became rather a broad term. In the late 19th century, Gilbert and Sullivan, the English creators of comic operettas, spoke of one of their subheroines as having "a caricature of a face." Perhaps, therefore, it is not surprising that, although cartoons as they are now known developed gradually out of caricature from the 15th century, cartoon is a 19th-century word.

Satirical deformations and comic analogies in sculpture, the drama, and vase painting are older than purely graphic caricature. The ancient Egyptians represented men as animals; Greek comedy had by-products in burlesqued figures on vases and in terra-cotta statuettes; Romanesque and Gothic sculptors made fun of human failings in stone capitals and wood miserere seat carvings all through the Middle Ages. The marginal flourishes of illuminated manuscripts contain grotesque faces and occasional exaggerated scenes from daily life, or references to the morality plays, which have the same relationship to those plays as Greek clay representations have to the stage. All such works verged upon caricature in the narrow personal sense; some were caricature in a broad sense. In the generations since caricature became a clearly defined idea, there have been occasional examples in painting and sculpture alongside the more usual drawing for reproduction.

Cartoon. A cartoon originally was and still is a drawing, a full-size pattern for execution in painting, tapestry, mosaic, or other form. The cartoon was the final stage in the series of drawn preparations for painting in traditional Renaissance studio practice. In the early 1840s, when that studio practice was rapidly decaying, cartoon rather suddenly acquired a new meaning: that of pictorial parody, almost invariably a multiply reproduced drawing, which by the devices of caricature, analogy, and ludicrous juxtaposition (frequently highlighted by written dialogue or commentary) sharpens the public view of a contemporary event, folkway, or political or social trend. It is normally humorous but may be positively savage. Just as the personal caricature was for an audience that knew the original, so the cartoon was and is based on wide acquaintance with the subject. It serves as a capsule version of editorial opinion when it makes political satire, and it is a running commentary on social change, sometimes intended as a corrective to social inertia. (For treatment of animated cartoons, see MOTION PICTURES: *Animation.*) (W.Am.)

Comic strip and comic book. The comic strip consists of a series of adjacent drawn images, usually arranged horizontally, that are designed to be read as a narrative or a chronological sequence. The story is usually original in this form. Words may be introduced within or near each image; they may also be dispensed with altogether. Words should never functionally dominate the image, which then would become merely illustration to a text. The comic strip is essentially a mass medium, printed in a magazine, newspaper, or book.

A comic book is a bound collection of strips, typically telling a single story or a series of different stories. Most of the better newspaper strips eventually appear also in book form.

Misleading use of "comics"

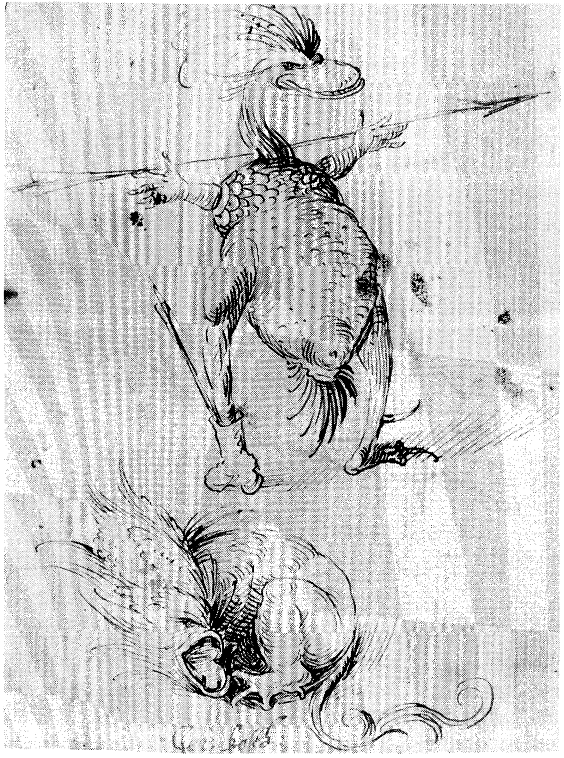
The word comic in connection with these strips is used only in the English language. Although now firmly established, it is misleading, for the early (pre-19th-century) strip was seldom comic either in form or content, and many contemporary strips are in no sense primarily humorous. The terms comics and comic strip became established around 1900 in the United States, when all strips were indeed comic. The French term is *bande dessinée* (*i.e.,* "drawn strip," or BD for short). The older German term is *Bildergeschichte* or *Bilderstreifen* ("picture story," "picture strip"), but the Germans now tend to employ the English word. In Italy the term is *fumetto* (meaning literally "little puff of smoke," after the balloon within which most modern strips enclose verbal dialogue). In Spanish the comic book is called *historieta.* (D.M.Ke.)

ORIGINS OF CARICATURE AND CARTOON

Individual satire. Caricature was a product of the Renaissance and Reformation emphasis on the importance of the individual. If a man was seen officially as an emperor, he was seen unofficially to have feet of clay or to be wearing no clothes. From about the first third of the 16th century the emphasis on decorum was so strong in Italy and spread so fast northward and westward, reinforced by a still more solemn decorum from Spain, that it produced a reaction. Desiderius Erasmus' *In Praise of Folly* was both a Renaissance effort at satire and a carryover of medieval mockeries; the marginal drawings made in one copy of it by early 16th-century members of the German-Swiss Holbein family are neither caricature nor cartoon in the modern sense, but they are in the same stream of subjective comment on objective observation as the series of exaggerated profiles drawn by Leonardo and Dürer. In the 16th century the work of the Flemish painter Pieter Bruegel the Elder is full of near caricature, as in the familiar drawing of an artist who is troubled at his easel by a nosy peering connoisseur behind his shoulder. In the work of Bruegel and that of his contemporary, the Dutch painter Hieronymus Bosch, there are witty and sometimes horrendous dislocations of parts of the body, combinations of human anatomy with fishes, birds, animals, and windmills, and exaggeration of obese or emaciated physical types that are likewise near cartoons.

Yet true caricature in the sense of the satirical portrait of an individual is almost impossible to identify before the work of a late 16th-century Bolognese painter, Agostino Carracci. The first caricatures of persons whose names are still known today are by Bernini. Artists of their time could at last afford to be easy about their position and occupation, about what was objective and what was subjective. They impartially caricatured one another in the studio and the types in the street. The rise of connoisseurship and collecting in the 17th century soon produced appreciation of these caricatures outside the studio, leading to the gathering of such drawings into albums and their transformation into multiple etched or engraved publications.

Staatliche Museen Preussischer Kulturbesitz, Kupferstichkabinett Berlin (West)



"Two Fabulous Animals," pen and brown ink drawing by Hieronymus Bosch (*c.* 1450–*c.* 1516). In the Staatliche Museen Preussischer Kulturbesitz, Kupferstichkabinett, West Berlin. 16.3 × 11.7 cm.

Development of the modern cartoon

Social satire. The less personal and caricature-like features of the modern cartoon developed slowly through the 16th century and up to the last third of the 17th. In part consciously based on Bosch and Bruegel, in part an autonomous protest against the Renaissance belief in order, symmetry, and fixed canons of beauty, there rose a European family of grotesqueries with sources as mixed as its appearance. Its only visible overall characteristic might be said to be that it reacted, from the depth of old folkways, against the novelty of the Renaissance, the New World, and the various hierarchies. Echoes of the dark Gothic forests mixed with fantastic reports of travel (some of them not really new but rehashed from Marco Polo) and with travesties of the Renaissance ornament of such works as Raphael's decorations for the loggias of the Vatican Palace. One result was the double images of faces and landscape or of human figures built of books, fish, or pots and pans by the 16th-century Italian painter Giuseppe Arcimboldo. In a slightly different area lay two types of pictorial comedy. The first type was the conscious satire by professional printmakers, illustrating such works as the *Ship of Fools* (1494), an allegory by the German poet Sebastian Brant, or issued as independent prints. Some of them were tendentious to the point of libel (such were Lucas Cranach's attacks on the papacy, inspired by Martin Luther); others were merely mockery, such as the 16th-century German painter Hans Beham's satirical view of mercenary soldiers and their camp followers. The second type was the unconscious pictorial comedy by comparatively untutored practitioners of woodcut or etching, hastily made and issued as newspaper extras might be issued now. These works were the ancestors of 18th- and 19th-century broadsides.

Class distinctions in caricature

If caricature deals with the individual and with what makes him individual, cartoons may be said to deal with groups and with their corporate characteristics; both are connected with the Renaissance love of classification and categorizing. Up to the time, in the mid-17th century, when both caricature and cartoon jelled in forms which answer to the modern definition, caricature had been prepared on the whole by knowledgeable artists; and cartoons had been prepared (though often with the assistance of professional artists) at a somewhat lower level. Thereafter, cartoons (in the modern sense of the term) came to be created in response not merely to artistic impulses but to the same sorting-out impulses that were creating the modern state and its society, its science, and its religion. Besides the sublime–ridiculous and the artist–layman dualisms in this field, there is a sort of class dualism that is a result of historical factors rather than of general human disposition. The history of caricature and cartoon is intimately connected with the history of the way in which one class has looked at another. Dürer looked down on the heavy-bodied man whom he called farmerish; yet within two generations Bruegel was drawing peasants in a sympathetic manner. A 16th-century Swiss engraver, Jost Amman, treated trades and professions dispassionately, as did the 17th-century French engraver Abraham Bosse, although Bosse did create some lampoons on unpopular types and callings. Jacques Callot, an early 17th-century French etcher, satirized dandies, beggars, and all classes between, and he even made the generally detested Gypsies (who were not yet romantic) almost sympathetic.

A rather comfortable social climate developed in the 17th century: the wars of religion had quieted in France and the Low Countries, though civil war occupied Britain in the middle of the century. But by the time the Restoration had quieted Britain, a general European threat appeared in the form of Louis XIV. Against his absolutism and against his influence through agents in Britain and the Netherlands, the first modern cartoon campaign was made. It was modern in that it was on a large scale—cartoons were published on a fairly regular serial basis, much like that of the modern daily newspaper's editorial cartoon (though the newspaper was still in the letter-and-gazette stage); and it was based on a fairly large and general acquaintance with the persons satirized if not always with the literary

"John Calvin," oil caricature by Giuseppe Arcimboldo, 1566. In Gripsholm Castle, Sweden. 64 × 51 cm.

By courtesy of Gripsholm Castle, Sweden

analogies used. The cartoons, which as yet made no particular use of personal caricature (that was still on a friendly studio footing), were prints made in Holland by a group of artists of whom Romeyn de Hooghe was the chief, and they were sold cheap. There had been Dutch political cartoons before, but they were laborious and appeared irregularly. The Dutch–English connection in the person of William III, the continuing threat of Louis XIV, and a succession of shattering events in various spheres stimulated a vast production of cartoons from the 1680s on.

Separation of cartoon and caricature

As the cartoon spread—along with its almost indispensable element, caricature—it began to divide; and although the periodical carrying cartoons was not to appear regularly until the third quarter of the 18th century, nor the true comic periodical until early in the 19th century, the divisions were real enough to follow from that time.

PERSONAL AND POLITICAL SATIRE (PURE CARICATURE)

18th century. About 1740 the English printmaker Arthur Pond published together 25 caricatures after original drawings by a number of artists. This collection must have been effective in spreading the idea and the word, for it was an excellent publication. Pier Leone Ghezzi, one of the artists included, was probably the first professional caricaturist, for he made a living with his pen portraits of Romans and visitors to Rome, many of which he engraved. He was a minor master in comparison to his contemporary the Venetian artist Giovanni Battista Tiepolo, yet the caricatures by the latter, delightful though they are, appear to deal with types or at least with anonymous individuals rather than with nameable ones. Tiepolo's pen-and-wash drawings were a small side issue in his enormous production, and they were not engraved or otherwise multiplied in his lifetime. Both artists had an eye for ungainly legs and posture and for odd clothing and the obvious features of the face.

On the heels of Pond's reproductions came handbill-like personal caricatures. Apparently begun in the 1760s by the Englishman George Townshend (later Marquess Townshend), these were comic portraits with punning titles or accessories, intended by disingenuous means to avoid being outright libellous. A flood of imitations followed. Soon Townshend's cards became comic illustrations in magazines such as *The London Magazine,* the *Political Register,* and the *Town and Country Magazine.*

Rowlandson's comic images

Early 19th century. *Great Britain.* Thomas Rowlandson created the comic images of a great many public characters in his day: royal dukes, actresses, auctioneers, hack writers of Grub Street. He was the perfect professional to carry on what the amateur Townshend had begun. Like Tiepolo, he was able to make the person and the costume assume a homogeneously ludicrous or pathetic–bathetic look, with factitious coiffures, wildly frogged uniforms, enormous bosoms and bottoms, and the dejected attitudes of trailing handkerchiefs.

Rowlandson's contemporary James Gillray was less of an artist and more of a professional cartoonist in the modern sense. Coming from the theatre to the political scene, he brought a highly dramatic sense of situation and analogy, but he was peculiarly violent and often scurrilous or scatological. His greatest talent was with patterns and ornaments of costume, which he would allow to take on a luxuriant life of their own. The Swiss-born English painter and teacher Henry Fuseli, though hardly a professional caricaturist, stood halfway between the painted Italianate caricature groups of Sir Joshua Reynolds, an academic English painter, and the Rowlandson–Gillray drawings and etchings; he had something of Gillray's theatrical manner, but his satirical drawings were more often sensual than scatological.

France. The French painter and engraver Philibert-Louis Debucourt might have equalled Rowlandson if he had not been so occupied with the intricacies of colour prints; but he produced a few superb cartoons of the Paris of his day, full of caricatures of fashionable personages.

The whole Napoleonic period gave rise to such passion that a school of French caricature was generated, the effect of which was to come after the Restoration. Two members of an older generation, Louis-Léopold Boilly and Jean-Baptiste Isabey, really began the work. Boilly, starting where Debucourt left off, satirized the modes and manners of the French. He was not in the direct sense a

By courtesy of the Allen R. Hite Art Institute, University of Louisville, gift of Janos Scholz, 1953

"Marchese Spada," pen-and-ink drawing by Pier Leone Ghezzi (1674–1755). In the Allen R. Hite Art Institute, University of Louisville, Kentucky. 20.3 × 29.2 cm.

"Amputation," etching by Thomas Rowlandson, 1785. 24.4 × 13.75 cm.

By courtesy of The Metropolitan Museum of Art, New York, Elisha Wittelsey Fund, 1959

political caricaturist but frequently used satirical portraits.

Isabey was primarily a portrait painter, but he was in contact with all the great continental political personalities of the first half of the 19th century, and he caricatured many of them privately. Most of the caricatures were not published, but they had an effect among artists, as did the satirical prints of Francisco de Goya.

The cartoon as political complaint

The tendency of the Restoration to suppress Napoleonic enthusiasms provided another rich ground for cartoon as political complaint. As soon as the first stage was over (in 1830), a change of administration was accompanied by the appearance of Charles Philipon's periodical *La Caricature,* the first great vehicle of Honoré Daumier, Henri Monnier, "Grandville" (J.-I.-I. Gérard), and others. The presiding genius had great politico-legal skill and knew exactly how far he and his artists could go. The famous likening of Louis Philippe to a pear, which was both a visual and a verbal pun, was not the least of *La Caricature's* successes. Daumier's colossal gifts included personal caricature, though in his later life he dealt almost entirely in more general social satire. In the early '30s he created for Philipon "Le Ventre legislatif" ("The Legislative Belly"), at once a political indictment, a rogue's gallery of caricatures, and a monumental composition; not to mention a long series of more detailed single antiportraits. Daumier's composite sociopolitical villain, Robert Macaire, and Monnier's Joseph Prudhomme, the sum of bourgeois pettiness, served as butts of satire when censorship caught up with directly personal caricature in Philipon's three or more overlapping papers.

Middle 19th century. *Great Britain.* The specifically cartoon-bearing journal was by this time an established fact. The *Monthly Sheet of Caricatures* had begun publication in London in 1830, lithographed like Philipon's journals. In these and other ventures, the publisher Thomas McLean issued hundreds of political caricatures during a great formative period of modern legislation; his artist, Robert Seymour, was in the Gillray line rather than the later one of John Doyle, who also worked for McLean. John Doyle's son Richard became one of the masters of the mid-century British school of subtler cartooning. The younger Doyle was one of the initial staff of *Punch* when Henry Mayhew started it in 1841.

First caricaturists of *Punch*

Punch began as a fiercely democratic weekly which applied to the young Queen Victoria and her growing family, as a matter of course, the same savage treatment that had been given by caricaturists of the previous hundred years to the Hanoverians. *Punch* was born in the years when the new Westminster Palace (Houses of Parliament) was under construction. Prince Albert and his team of artistic advisers, wishing to revive fresco painting as a means of decorating the vast wall spaces of the new buildings, opened a competition. When the cartoons (in the original sense of the word) of the proposed compositions were exhibited, it was evident that many of the painters were unfamiliar with work on so heroic a scale, and some of their designs were ludicrous. *Punch* was quick to satirize them in a series of "*Punch*'s Cartoons." One of the reasons for "*Punch*'s Cartoons" on the Westminster hall exhibition was that the unpopular foreign Prince Albert had fomented the competition. But *Punch* was not so much merely antiroyal as it was antihumbug (the novelist W.M. Thackeray, after all, was another of the staff). Though its first few years were marked by some of the worst puns in a punning age, it not only kept up, with its fairly small fixed staff, a remarkable fire of lampoon but soon developed the weekly full-page political cartoon, certainly one of the chief ancestors of the modern editorial cartoon. Nineteenth-century British and general European politics might be briefly comprehended in a few dozen of these. The most famous is probably still that farewell to Otto von Bismarck, "Dropping the Pilot," by Sir John Tenniel.

The woodcut technique used for many decades in *Punch* caused the cartoons on the political page, the largest in scale of the cartoons, to be almost invariably dull in surface. The speed required of the divided-labour teams which produced the cuts from the artists' drawings did not allow for subtleties.

Other countries. The earliest really impressive makers of personal caricature and political cartoon in the United States were David Claypoole Johnston and Thomas Nast. Nast first made his name with American Civil War cartoons in *Harper's Weekly,* which like *Punch* used the woodcut process with an elaborate division of labour in the back shop for the rapid reproduction of cartoons.

The middle and late century produced in Italy the brilliant political caricaturist Virginio, who was to the rise of Italian nationalism what Nast had been to the North in the American Civil War; he worked for *Il Fischietto* of Turin. In 1848 *Kladderadatsch* started in Berlin. Munich had *Fliegende Blätter* and *Punsch. Punsch* was more political than the others, which were long-lived comic weeklies in the social-comment style. J. C. Schleich's *Punsch* cartoons were a running Bavarian comment on Prussianism.

Nast's personal satire

Late 19th century. *The United States.* In the boom days of the 1870s, Nast became a master of personal satire; his long practice in dealing with the professional wood engravers gave him at last a style and scale that triumphed over finicky crosshatching and gave full effect to his ruinous attacks on William Marcy "Boss" Tweed and other grafters; he was one of the most fertile of symbol makers outside the personal field and is probably the creator of the donkey of the Democratic Party as well as of the Tammany Hall tiger of the New York Democratic organization.

In 1876 *Puck* was founded. It was soon to develop new artists, notably Joseph Keppler and Bernhard Gillam. They worked in a lithographic style of considerable artistic competence, without the force of Nast or the effortless flow of Daumier, but with plenty of clever analogies and with an understanding of the sort of likeness required in caricature.

Harper & Row, New York

"Who Stole the People's Money?—Do Tell. 'Twas Him," by Thomas Nast (1840–1902).

England. *Punch* meanwhile had settled into its richest period, with Tenniel and Harry Furniss as political cartoonists. *Vanity Fair* (from 1868) offered some competition, especially at first with its regular coloured lithographic antiportraits. These were signed "Ape" (Carlo Pellegrini) and "Spy" (Leslie Ward, later knighted); they kept up a steady supply of big-headed comic figures against an almost invariably blank background. They also kept up the old device of never quite naming the subject in so many words, but as they were directed at a public which was "in the know," this was part of the fun. These colour caricatures were much loved, and were often framed and hung on private walls. Max Beerbohm (knighted in 1939) devoted himself largely to social and literary satire but almost always on a basis of personal caricature. His deceptively understated outlines and pallid washes, the latter used as local colour for the sake of the overall design, were the perfect means for parodying the good taste of the *fin de siècle.* His symbols for the writers G.K. Chesterton, G.B. Shaw, Joseph Conrad, and W. Somerset Maugham have become almost the standard views of those writers.

Influence of photomechanical reproduction

Technical developments. Toward the end of the century there was a rebirth of personal satire which accompanied new techniques of reproduction and perfected enrichments of such older techniques as colour lithography. Photomechanical reproduction, especially after the development of halftone, allowed direct reproduction of the artist's drawing without personal interpretation by wood engravers or other technicians. Colour lithography, which had been either limp or turgid on the whole, found a new life. The caricaturists who had been able to draw directly on the stone, as Daumier did, had always had more freedom and better control over results than those who worked with pen and paper; now the latter could depend upon themselves and a photomechanical process.

France. Henri de Toulouse-Lautrec produced large-scale posters and, earlier, polychrome lithographs for the Parisian publication *Le Rire* (from 1894) and for independent distribution. He created a new style of informal composition, somewhat influenced by Japanese prints, with bright clear colour, broad, rather casual outlines drawn largely with the brush, a trick of making tone by means of spatter, and a wit that saw through ugliness to a new sort of eloquence. His view of Oscar Wilde was economical and devastating, and his caricatures of theatre and music hall personalities are unmatched. Another French artist, Caran d'Ache (Emmanuel Poiré), worked on a smaller scale with pen and brush and was one of the most effective continental commentators on the South African War.

Germany. The artists of the Munich satirical publication *Simplicissimus* (from 1896) were all somewhat influenced by Toulouse-Lautrec in their use of white space, spatters, and often random outline; they all commented on those features of German life that were most disliked outside Germany—the didactic professor, the tourist, and the military dandy. Their caricatures in the last field were very thinly veiled; Eduard Thöny, one of this group, was especially popular for the way he conveyed the upper-class boorishness of Prussian officers.

20th century. *The United States.* Charles Dana Gibson was a virtuoso of the pen, using the manner of *Punch*'s Phil May as a point of departure. He used the pen as he pleased, sometimes in a direct descriptive manner, sometimes with colouristic suggestion, sometimes almost antigraphically. Though he helped to create the caricature types of Theodore Roosevelt and Woodrow Wilson, he was more a social than a personal caricaturist. He is mentioned chiefly because he introduced American physical types called Gibson girls and Gibson men.

Role of *The Masses*

A position much like that of *Simplicissimus* was occupied in the years 1911–17 by *The Masses* of New York, which had an editorial policy based on socialist idealism. It was served by a group of artists whose fine drawing made their often sharp propaganda tolerable in quarters where it might not otherwise have gotten a hearing. John French Sloan, George Bellows, Boardman Robinson, and Art Young were as likely to deal in social terms as in personal ones, for by this time personal caricature was moving into the newspaper editorial cartoon or the pages of theatrical or sporting news.

Photomechanical reproduction not only allowed greater freedom for comic artists; it made possible the daily newspaper cartoon and later the syndicated editorial cartoon and the comic strip. About the same time as the new generation of weeklies there was a rise in the use, the autographic character, and the influence of pictorial journalism. John T. McCutcheon of the *Chicago Tribune,* though he used a rather dry and old-fashioned pen technique, was able to range over politics, the "good old days," the mores of the moment, and sports. His cartoon world, like that of *The Masses,* was almost entirely urban, but he was one of the first of a generally imperturbable type of American

"Tricks or Treaties,"—from *Straight Herblock* (Simon & Schuster, 1964)

"Tricks or Treaties," by Herbert Block.

cartoonist, whose view is amused rather than aroused. His career ran from before Theodore Roosevelt to after Franklin D. Roosevelt's time. In his line of succession stood such men as Edwin Marcus and S.J. Woolf in *The New York Times,* Oscar Cesare of *The Sun,* Herbert Block ("Herblock") of *The Washington Post,* Daniel Fitzpatrick of the *St. Louis Post-Dispatch,* Rollin Kirby of the *New York World-Telegram,* Bill Mauldin of the *Chicago Sun-Times,* John Fischetti of the *Chicago Daily News,* and others too numerous to list.

Dominance of David Low

England and the Continent. The outstanding political commentator of the first half of the 20th century was David Low, a New Zealander, who worked for the *Sydney Bulletin* before going to Britain. Low was perhaps the best all-around man in the field of caricature since Daumier. His brush drawing was of an Oriental economy, his invention of analogy gleeful without being really outside the classic British educated tradition, and his hatred reserved for a few needful occasions. Like many before him, he employed hackneyed devices (*e.g.,* the heads of a pack of British politicians on dogs' bodies) but by slyness of expression always managed an original twist. There was almost no one in the political field to touch Low except for the Dutchman Louis Raemaekers during World War I, and Raemaekers was bitter where Low was dry and crisp, with footnotes of rumbling laughter. Jean-Jacques Sennep (pen name of J.-J.-C. Pennès) of Paris' *Le Figaro* and Fritz Meinhard of the *Stuttgarter Zeitung* were important French and German caricaturists of the 20th century.

COMEDIES OF MANNERS (THE CARTOON)

Types and groups, rather than politics and the politician or any nameable individuals, are the concern of the comedian of manners. He may love mankind for its imperfections or set out to seek improvement, but his method will be much the same. He does not need, as the political cartoonist does, to set up allegories and analogies or to write names on labels, but he may sometimes sharpen his comment by treating human beings as animals (monkeys and apes for obvious reasons have long been the favourites, along with dogs and birds). If the personal caricature is an antiportrait, the cartoon on human foibles is often a sort of anti-sumptuary law or a countergrammar which says, "The exceptions are more fun than the rules."

16th to 18th centuries. Bruegel and Callot were certainly comedians of manners. Bruegel's picturizing of Flemish proverbs, themselves often comments on foibles, and his prints of the Seven Deadly Sins with satirical examples filling the backgrounds combine a bit of moralizing with the delighted empathy of a participant. Callot is slightly more detached, possibly because of his more conscious style and because he was himself the printmaker (Bruegel drew for professional engravers and woodcutters); but in the catalogs of byplay in his panoramic scenes of fairs and in his trick of making the beggar wear his rags handsomely, he is always balancing and measuring.

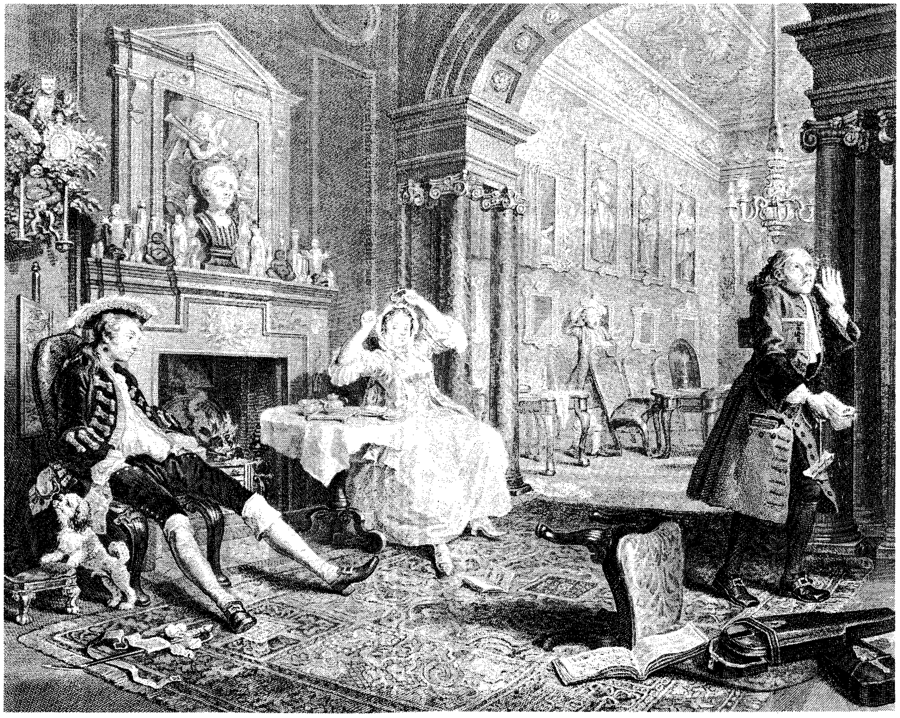
Parallel to two-dimensional comment in this vein ran the theatre, notably the commedia dell'arte, puppet theatre, and the performances of the jester and clown. Both appeared in the late 17th and early 18th centuries in the rather courtly comic drawings of the Frenchman Claude Gillot, Antoine Watteau's predecessor. These are not really comedies of manners, for the clowns are used as if they were monkeys aping human ways at a remove toward greater elegance rather than toward apishness. They point the way to a good many 18th-century practices: Gillot and Watteau both made decorations that included monkeys (singerie) just as Francois Boucher and later artists were to use pseudo-Chinese scenes (chinoiserie) occasionally as ways of commenting on contemporary European life.

Hogarth's cartoons of manners

It is with William Hogarth that the cartoon of manners reached great stature. His series "Marriage à la Mode," "A Rake's Progress," "A Harlot's Progress," the "Four Stages of Cruelty," and the unfinished "Industrious and Idle Apprentices" were loaded with observation not only of human beings but of objects and their ecology, as if he were using his own proliferation of comic images in protest against waste of time, talents, life, and pride. Hogarth, like Sir Joshua Reynolds after him, even painted comic subjects, but he kept to social satire and avoided personal caricature. His pictures of depravity and ferocity are hard to beat, but he could put an expression of by no means unholy delight on a wicked face. In the "Laughing Audience" he gave a full measure of laughter. Hogarth's engravings ran to very large editions and were recut and reissued and then copied at reduced scale for books of the *Complete Works.*

19th century. *Spain.* Francisco de Goya is hard to place in the historical development of the comedy of manners. His "Caprichos" (1796–98), etchings prepared by some of the most simple and trenchant brush drawings ever made, appeared in the last years of the 18th century and

By courtesy of The Metropolitan Museum of Art, New York, Harris Brisbane Dick Fund, 1932



"Marriage à-la-Mode: Breakfast Scene," engraving by William Hogarth, 1745. 35.2 × 43.8 cm.

The folk wisdom of Goya

can be called comedies of manners only insofar as they are related to folk sayings and the bittersweet Spanish folk wisdom. Thus, they stand in the line of Bosch and Bruegel, so many of whose paintings were in Habsburg collections in Madrid. The "Proverbios" of 1813–19 are even more monumental transfigurations of various states of the human condition. Like the "Caprichos," they used the caricaturist's means for irony and satire, but there was little of the comic left in them and none at all in the "Desastres de la guerra" (1810–14, "Disasters of War"), which used the Peninsular phase of the Napoleonic Wars as a point of departure. They are closer to universality than even Callot's similarly inspired series and are searching comments on more stages of cruelty than Hogarth covered. In them, Goya was really a political cartoonist using no names; yet he was hardly a public cartoonist in the normal sense because censorship and other factors allowed only a very small circulation of his later work until a sizable edition was printed a generation after his death. The earlier work, which contains elements of comedy, did get abroad and had influence in France and England probably before Goya's death. Artistically, if not politically, his work would have had the same powerful effect whenever "discovered" or circulated.

France. "Paul Gavarni" (Sulpice-Guillaume Chevalier) was more purely a comedian of manners than Daumier, though he was no less perceptive and no less sympathetic with the *petit parisien.* He had a grace derived from his apprenticeship in fashion illustration that produced enchanting jokes on young people in love, dandies, and the theatre and circus. He worked late in life for the *Illustrated London News,* as did Constantin Guys, the French foreign correspondent who reported the Crimean War to the British. Guys, a prolific draftsman who always kept a comic touch, was peculiarly subtle in reporting the great but contrived elegance of Napoleon III's court. He helped both British and French to see themselves as others saw them. "Grandville" was a comic artist on *La Caricature* whose work recalls some of the complicated inventions of Arcimboldo.

Daumier was, of course, the great master of social comedy with or without political content. His series of affectionate if disenchanted comments on married life, the theatre, the courts, concierges, musicians, painters, bluestockings, bathhouses, and children constitute as full a report on Paris in his time as Rembrandt's drawings were for 17th-century Amsterdam. The words were often important, especially when Daumier was indicating in his text the unspoken thoughts of his characters (thus anticipating the 20th-century cartoon in which a thought or vision is indicated as a balloon with cloud-scalloped edges and a picture rather than words inside). His often untidy line and knowingly casual accents of tone produced (at will) sensations of chill weather, of ecstasies of gluttony, of juvenile pride, or of legal craftiness.

Courtesy of the Fogg Art Museum, Harvard University

"Actualities No. 190; Des dames d'un demi-monde mais n'ayant pas de demi-jupes" ("Demi-mondaines with more than demi-skirts"), lithograph by Honoré Daumier, 1855. 19.7 × 25.7 cm.

England. Rowlandson, as noted earlier, was a political caricaturist part of the time, but above all he was a lampooner of ludicrous and excessive behaviour. He created almost unaided a gallery of types missed by Hogarth, many of which persist in British life—the antiquarian, the old maid, the harried foreign servant, the pleasantly blowzy barmaid, the decent old parson. He was by no means as bawdy as he is supposed to have been, but he liked to push action, like appearance, to an extreme. His *Dr. Syntax* may be called an ancestor of comic strips.

George Cruikshank carried Rowlandson's methods almost beyond extremes in his youth. He used superfantastic costume and sometimes that device of enormous heads which some 17th-century caricaturists used and which is still used by sports-page cartoonists and comic advertising artists.

Humorists of *Punch*

The longest continuing habit and tradition of humorous comment on the passing world has been made by the English humour magazine *Punch.* Though it began in puns and peevishness, it warmed up during the 19th century with John Leech, Charles Keene, George Du Maurier, and in the 20th century with George Belcher, "Fougasse" (Kenneth Bird), H.M. Bateman, Nicolas Bentley, E.H. Shepard, and Osbert Lancaster. Leech was in a sense the pictorial equivalent of Thackeray (Thackeray was an excellent comic draftsman but better at getting the feel of past time with a comic flavour than at considering his contemporaries other than in words). Leech and Keene belong to the era of wood-engraved reproduction; when one sees their original drawings and manuscript captions or dialogue, it is apparent that something was lost in detail and finesse of line but nothing in sense of comedy, in the affectionate tone. The enormous self-confidence of the optimistic Victorians, expressed at first through the violent or bumptious Regency manner of the young Cruikshank, was tempered by the staff-meeting or meeting-of-minds conduct of *Punch.* The "manners" part of the phrase "comedy of manners" became subjective as well as objective. *Punch* became an upper-class weekly and continued as such for three or four generations, reflecting the large knowledge of all classes it was possible for its staff to offer its readers, and the large delight of the upper class in seeing its own foibles and those of its servants, tradesmen, lame ducks, and "climbers" exposed. The swing of a crinoline by Leech and the curl of a cabdriver's hat brim by Keene were perfect selective imitation, themselves almost inimitable; the crinoline and the hat are gone with those who knew how to wear them, but the picture in *Punch* remains.

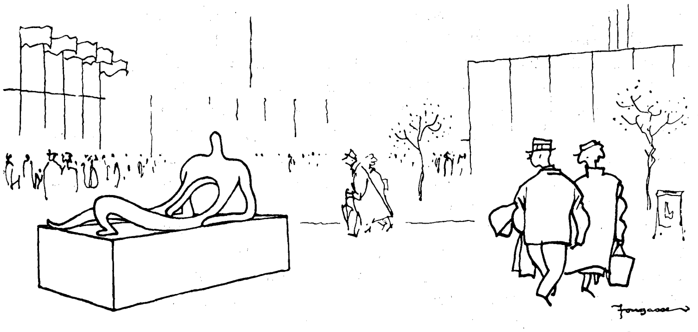
Photomechanical reproduction came in during Du Maurier's day but hardly affected his generation of artists. May's pen was better served by the camera and the zinc block than Leech's had been by end-grain boxwood and gravers, but the general language was the same. With the generation of Belcher there was a great change. His own crumbly charcoal or crayon strokes were perfectly adapted to the new process—indeed it was mutual—as were the fat black lines of the Frenchman Jean-Louis Forain and the mid-20th-century cobwebs of Rowland Emett. Fougasse's highly personal little curly stick men, drawn perhaps with a signwriter's pen, could be reproduced by almost any method, but the sharp lines and solid black areas of Bateman (deriving ultimately from Aubrey Beardsley's decorative style) and the thick–thin pen strokes of Shepard (more in the Keene tradition) were well served by modern processes. Shepard was more truly an illustrator than a cartoonist, but Bateman's towering humours and bulging-eyed apoplectic businessmen were in the direct line from Edward Lear to such frantic American cartoonists as Virgil Partch.

Contributions of Edward Lear

Lear practiced as a comic draftsman an economy and geniality that are hard to improve upon, but like Daumier he supposed that his own best gifts lay in another field. Humour had been brought into satire by Hogarth; a truly funny style of drawing was brought into cartooning by Lear. Hitherto, standard drawing techniques had been applied to grotesque shapes and comic situations, but Lear's line went wandering off into a sort of joke on calligraphy. Furthermore, he travelled into areas of fantasy previously barely hinted at.

Aubrey Beardsley used a caricaturist's methods, but little of his work, except perhaps the illustrations to *The Rape of the Lock,* was actually in intent or effect caricature or cartoon. If some of Cranach's prints can be called illustrated libels, some of Beardsley's can be called illustrated yearnings by unfrocked lechers. They are important in this connection because their combination of large white spaces, clear lines, and solid or slightly irritated blacks could be reproduced successfully in a choice of dimensions and thus laid down a discipline for illustration, commercial art, and the comic strip.

© *Punch,* London



"That reminds me, dear—did you remember the sandwiches?" by Fougasse (Kenneth Bird).

20th century. The whole tenor of pictorial comedy was shifted by World War I and by the boom times thereafter. Some previously forbidden subjects became admissible. Political caricature during and after the war was excessively partisan, while the cartoons about the war itself tended to alleviate the pain of the struggle. Bruce Bairnsfather's Old Bill and his colleagues in Britain got through it by joking. After it was over, the public for comic publications was greatly enlarged; while the newly rich were standard butts for cartoonists catering to all classes, they were themselves buying comic weeklies.

In the United States the usually monthly comic magazines of universities and colleges had a sudden flowering, to such an extent that an anthology for their cartoons called *College Humor* was published for several years in the 1920s and '30s. The tendency of previously serious weeklies to use small cartoons here and there or to insert a funny page somewhere created not only new markets for cartoonists but also a temporary decline in the purely humorous magazines, and *Punch, Life,* and *Judge* had difficulty surviving the Depression of the 1930s. The title of *Puck* had already passed to a newspaper chain which used it for a Sunday supplement, and the title of *Life* eventually passed to a periodical of different character. *Simplicissimus* never quite rose between wars to its pre-1914 stature; the effort which went into Dada and Surrealist publications in Germany and France in the '20s, when art itself became an object of social satire, meant a loss to other comic publications. Meanwhile, the public in general gradually became aware of modern art, and its presumed incomprehensibility became almost as routine a subject by 1940 as mothers-in-law or freshly painted park benches.

In the United States an older generation of humorists somewhat of the upper-class *Punch* style lingered briefly after World War I. Of such were Oliver Herford, whose *Alphabet of Celebrities* and other comic verses with pictures were published as small books; Peter Newell, whose highly original *Slant Book, Hole Book,* etc., had a sharp eye to late prewar costume, and Gelett Burgess, whose *Goops* for children were spaghetti-like little figures whose behaviour illustrated a moral.

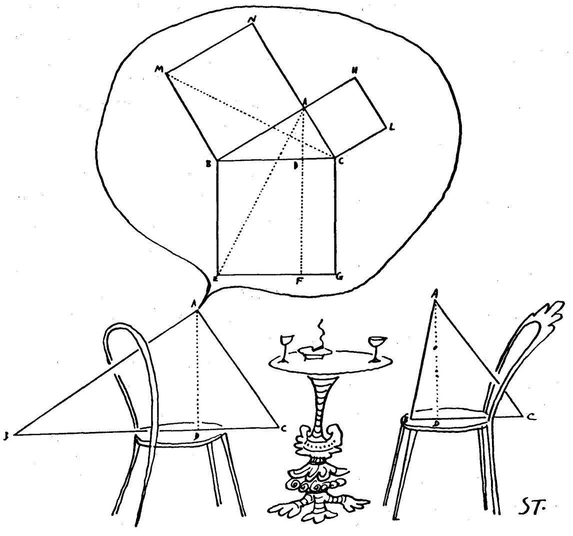
But to these was now added a new generation of sophisticated but slightly flashier performers, many of them with theatrical connections, many at first employed by the fashion magazine *Vanity Fair* and later by *The New Yorker* (beginning in 1925); Ralph Barton, who did superb roués; Rea Irvin of the thin trembly line, poached eyes, and almost oriental splendour; John Held, Jr., whose angular young women helped define the Jazz Age; Gluyas Williams and Ellison Hoover, who satirized business, industrial labour, and other subjects not well known to the *Punch* tradition; and Alfred Frueh, whose caricatures of theatre people recalled Toulouse-Lautrec. The Depression of the 1930s brought forward a few artists with a genius for social protest, few of whom had any real sense of comedy because tragedy was not to them, as it had been to Daumier, the other side of the same coin. In the United States the Communist *Daily Worker* had the services of William Gropper, a distinguished lithographer and editorial cartoonist who was sometimes able to capture something of the humorous tone of the prewar *Masses.* And it gradually became known that in the Soviet Union a comic magazine called *Krokodil* was allowed to gibe at the ways of its brothers and even occasionally of its masters.

Cartoons of the Depression years

The two most interesting features of cartoon and caricature in the first half of the 20th century were the rise of the one-line joke and of the pictorial joke without words, and the enormous diversity of styles of drawing. *The New Yorker* was probably the inventor or reinventor of the one-line joke and certainly its chief fomenter. Five-decker dialogues with headings were swept away even from *Punch,* and there was a greater unity of words with picture, paralleling the tendency toward tabloid newspapers with large photographic halftones and very pithy text. The joke without words, often in two or more frames, was the extreme of economy of language. One result of this change was that the comedy-of-manners cartoon must convey its comment entirely through costume, setting, and (to a lesser extent) situation, and the emphasis thus tended to fall more on comic situation than on plays on words, class differences, or marked action. *The New Yorker* and magazines whose cartoons had been influenced by it aimed at a sophisticated audience. *The New Yorker* itself, while enjoying in its maturity a position equivalent to that of *Punch* in the 1880s, aimed its advertising and much of its writing at upper-income classes, but its cartoons were aimed at the classes described as highbrow and upper-middle-brow. Such features of the old-fashioned British-style upper class as servants were always treated by *New Yorker* cartoonists (notably Mary Petty) as necessarily comic fossils of an old order and hence in, rather than out of, that old upper class. Whole new areas of social-comedy subject matter arose in this magazine: the life of the Jewish community, the fauna of bars, the managerial class and its flavour, the lighter side of the well-kept woman, commercialized sports, the imagined life of colonies or races of antisocial beings, and what might be termed the comedy of the upwardly mobile.

Drawing by Chas. Addams; copyright © 1949 The New Yorker Magazine, Inc.

"For heaven's sake, can't you do anything right?" by Charles Addams.



Captionless cartoon by Saul Steinberg.

© 1961 Saul Steinberg from *The New World* (Harper & Row). Originally in *The New Yorker*

The diversity of styles of drawing reflected the influence of Postimpressionist art quite as much as did the use of modern art as a subject for jokes. The great draftsmen who were on the edges of Impressionism (such as Toulouse-Lautrec) had much influence on caricature and cartoon; while the same photomechanical reproduction that advanced the latter communicated modern painting to a vast public. The loose, almost deliberately ugly method of the Expressionists got into some of the single-cartoon commentators to such an extent that their shorthand was sometimes difficult for those who did not read them daily. The meandering willful line of the 20th-century Swiss painter Paul Klee certainly influenced Saul Steinberg; the Cubists' studies in African sculpture were echoed in cartoons by Miguel Covarrubias and Partch; the "classical" period of Pablo Picasso in Richard Taylor and others; the curving economical line of Henri Matisse (oddly enough) in Richard Decker. Occasionally, cartoonists parodied one another: Oliver Herford once presented Gibson girls as paper dolls without expression; Al Capp in the comic strip "Li'l Abner" parodied "Dick Tracy" and "Mary Worth." *Mad* magazine parodied everybody: style, subject, everything from politics to pornography.

The cartoon of predicament or situation

The comic strip having taken over the comic presentation of events almost completely, especially since the rise of the one-line joke or picture caption, the modern cartoon became one largely of situation or of predicament that was stated without a solution's being worked out or suggested. The mother of a sideshow circus family, confronted with more than she can manage, simply says that after all she has only three hands; a young girl is in ecstasies over a sunset, while behind her a bearded artist-stereotype says, "Too much purple." Meanwhile, the longer sort of comic anecdote retreated to a purely oral–aural life or to the bound volume of jokes, where it sometimes had a vignette-like illustration.

The cartoon of situation was certainly not new, but it predominated in the first half of the 20th century. A Daumier lithograph showing a very fat woman in a crinoline climbing into an omnibus bore no dialogue, but simply the caption, "A mere nothing, and the bus is full." This was a cartoon of predicament. There has tended to be a cluster of these situation subjects: the desert island no larger than a hearthrug, the man who meets a woman walking and (in a scalloped balloon) imagines her naked, the flying carpet with novel chauffeur or passenger, the picture gallery with mutual reaction between work of art and viewer, the psychoanalyst's couch, the big-game hunter's trophy room. If the situation was clear, not even one line of joke was required. Such cartoons had a sort of family connection with the earliest caricatures, but they were not merely antiportraits of types, they were portraits with accessories that created the predicament. So were the tiny single woodcut figures that were inserted as pictorial puns into the text pages of *Punch* in the 1840s. But the latter-day predicament may be highly complicated; in the hands of such a cartoonist as George Price, whose split pen line built up tattered edifices of dowdiness, or Emett, whose fantastic locomotives and wispy codgers were half infernal and half heavenly, the comedy came from an accumulation of frustrating but ludicrous detail. Frustration, that renowned companion of modern life, was dissolved by laughter. Even the presumably invincible American businessman was often represented in cartoons in frustrating situations, often briskly indicated by the graphic lines on the charts in his office (Whitney Darrow excelled in this genre). André François, who worked for both French and British papers, was a master of the rapidly sketched situation; so was "Anton" of *Punch* (a man and a woman jointly using the name), who kept up the tradition of satire through clothes, being particularly good at pseudo-Edwardian nattiness. Herb Stansbury's "Smart Chart," a one-frame comic for the financial page, satirized stock market graphs. In the drawings of the Romanian-born Saul Steinberg there was almost a parody of the cartoon of situation, for his lines doubled back on themselves and bit their own tails: the hand was indicated as drawing the portrait of which it was a part, or the frustrating details positively engulfed the subject (a wicker chair taken over entirely by its curlicues; tattooing extended beyond the tattooed man; the woman with a lozenge-shaped face, on her lap a baby whose lozenge-shaped face is one-quarter of the larger lozenge).

Unique style of Steinberg

Yet there were also extraordinarily simple performances; the unassuming little people of Jean Effel (François Lejeune) moved gently through the trials of Adam and Eve; Jacques Faizant's bad children produced hilarious effects by conveying their concentration in a few lines; Otto Soglow's stenographic vocabulary of forms for human bodies (perhaps slightly indebted to Burgess' *Goops*) was so graphic that it could be used in minuscule dimensions with perfect legibility. On the other hand, Peter Arno's large-scale and heavy outlines, despite simple straightforward design, made his beaky and bosomy figures almost jump off the page (for many years one of his near-bawdy cartoons almost invariably occupied a position in *The New Yorker* on the full page immediately after "The Talk of the Town," which suggested that the political cartoon of *Punch* was being ridiculed). Ronald Searle, after a long British career of making spiky and raffish pseudo-Edwardians and fiendish schoolgirls, had a success as an artist for American advertising. A pair of delightful opposite numbers were W. Heath Robinson and the slightly later Rube Goldberg, who on both sides of the Atlantic created wild half-anthropomorphic parodies of intricate machinery. During World War II, Bill Mauldin's disenchanted soldiers were proper descendants of Bairnsfather's. (W.Am.)

COMIC STRIPS AND COMIC BOOKS

Early strips. The comic strip, defined as a mass medium, cannot reasonably be said to exist before the invention of printing. In the early period two principal forms may be observed: the series of small images printed on a single piece of paper (narrative strip proper), and the series composed of several sheets of paper, with one image per page, which when displayed on the wall of a house formed a narrative frieze or picture story.

Early German woodcuts

From the outset two basic groups of themes emerge: political morality and private morality. Surviving pre-1550 strips, most of which are German woodcuts, deal with such subjects as the lives of saints (subdivided in the manner of late medieval painted altarpieces, an arrangement that may be considered a decisive factor on the compartmentalization layout of broadsheets); accounts of contemporary miracles; mockery of worldly love; and politically inspired accusations against the Jews. The Reformation and the ensuing wars of religion through the 17th century, particularly in Protestant Germany and the Netherlands, gave rise to many propagandistic and patriotic strips based on contemporary political events. In the course of the 17th century the narrative strip, hitherto an ill-defined and irregular phenomenon, became stabilized and typically took the form of an allegorical centerpiece

"Profanation of the Host by Jews at Passau 1477." German broadsheet, *c.* 1490.

surrounded by narrative border strips. Although generally crude in style, these strips manage to render accounts of political intrigue and moving descriptions of military terror; the best known in the latter category is the exquisitely executed and carefully cadenced narrative of the Thirty Years' War by Callot. Little known, but as powerful in their way, are De Hooghe's indictments of Huguenot persecution under Louis XIV. Romeyn, the first named artist to devote himself consistently to the narrative strip, also left colourful, forceful, and elaborate accounts of the accession of William III to power in Holland and England. English engravers, inspired by the Dutch example and led by Francis Barlow, retailed the complex political events of the period in the form of playing cards, which were often sold in uncut broadsheets.

The earliest strips concerning private morality are German and recount atrocious forms of murder and their public punishment, the emphasis shifting from the latter (in the 16th century) to the former (in the 18th century). The crime strip eventually developed into the more or less exaggerated and romanticized life of the famous brigand, which is the precursor of the modern detective strip. Narratives based on a wider spectrum of ideally immoral behaviour took as their point of departure illustrations for the parable of the Prodigal Son, woodcut versions of which, independent of the biblical text, were first produced by Cornelis Anthonisz of Amsterdam. The riotous living of the Prodigal, enriched with elements from illustrations for the Seven Deadly Sins and the Ten Commandments, was distilled in various Italian lives of harlots and rakes, the most comprehensive and drastic of which are mid-17th century Venetian. A generation later, the Bolognese artist G.M. Mitelli was giving his narrative and semi-narrative satires almost caricatural moral emphasis. German artists in the 17th century specialized in satirically exposing the tyranny of shrewish wives and proposing violent remedies. The Dutch at this time produced frankly farcical strips of primitive design expressly for children. By mid-18th century the Russians, too, were making satirical strips.

Hogarth's picture stories

The various social and moral themes that had been crudely treated in different countries and at different times were the raw material for the English artist Hogarth, who raised the anonymous broadsheet picture story to an aesthetic level which rarely has been surpassed. With a social insight both broad and deep, with an unrivalled sense of satirical counterpoint and topicality of reference, and with exceptional physiognomic finesse, Hogarth dealt with types from all classes of society. His narrative richness is entirely visual, for he dispensed with all the broadsheet paraphernalia of caption–balloon–legend–commentary, permitting only such inscriptions as can be introduced naturalistically into the scene. Hogarth's moral attitude was also new: he depicted the follies and the punishment of his protagonists with a measure of sympathy, reserving the full fire of his satire for those who exploit them. Among Hogarth's many followers, two stand out: the German Daniel Chodowiecki, who reduced the Hogarthian picture story within the compass of almanac illustrations, and the Englishman James Northcote, who tried to combine Hogarthian realism with a Neoclassical sentimentality ("Diligence and Dissipation," 1796).

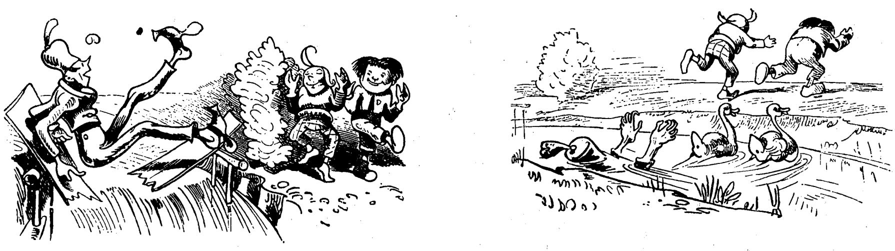
It was the introduction into the broadsheet of the essentially comic mechanism of caricature that established the "comic strip" as basically comic in both form and content. The major exponents of the caricatural strip during the "Great Age" of English caricature around 1800 were minor artists such as Henry Bunbury, George Woodward, and notably Richard Newton, who in his brief career combined elements of Hogarthian satire with the grotesque exaggerations of Rowlandson and Gillray. Economy of line, instantaneousness of comic effect, and visual and verbal wit now became the hallmark of the strip. With the story concentrated on a single page, backgrounds and narrative incident were minimized in favour of striking facial expressions and silhouetted poses.

Rodolphe Töpffer

19th century. The heir to the experiments of the English caricaturists, and the father of the comic strip in its modern sense, was Rodolphe Töpffer, a schoolmaster of Geneva, whose life spanned roughly the first half of the 19th century. Exempt from the personal–political preoccupations of the English caricaturists, Töpffer created a species of absurdist antiheroes who struggled desperately, fruitlessly, and farcically against the caprices of fate, nature, and an irrational, mechanistic society. The stories (lithographed in little oblong albums containing up to 100 pages) are purposefully purposeless, flow with calculated non sequiturs, and make digression a narrative principle. The pace is sustained by another revolution in draftsmanship, for Töpffer discovered how to turn systematic doodling to account, how to exploit the accident, and how to vary physiognomies experimentally. By abandoning anatomical, three-dimensional drawing, he showed how to render movement for movement's sake. Töpffer's strips are also morally mobile: the normal relationship between cause and effect or crime and punishment, which had underpinned all the older stories, disintegrated.

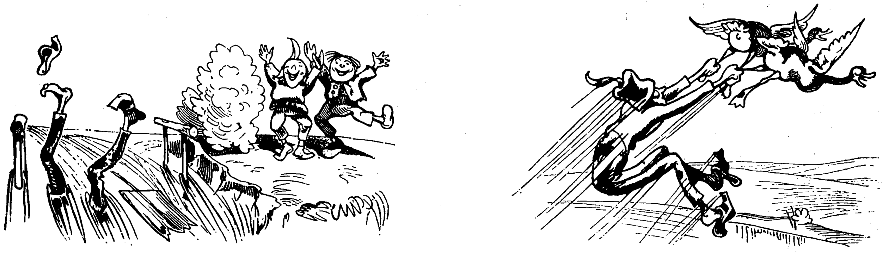
The Töpfferian mode was introduced into the new French satirical periodical journalism by various minor artists. By entering the regular newspaper press, the old broadsheet cartoon exchanged independence for security and became, for better or worse, subject to editorial control. Cham (pseudonym of Amédée de Noé) published several albums modelled on Töpffer in the 1840s, before settling for a style nearer to that of Daumier. The photographer Nadar (pseudonym of Gaspard-Félix Tournachon), among others, flirted intermittently with the new mode, rendering style and story line more realistic and therefore heavier. Léonce Petit, who had Töpffer's lightness of touch but not his imaginative flair, specialized in rustic farce. A special place is occupied by the illustrator Gustave Doré, who developed a new method owing relatively little to Töpffer (shown in, *e.g., Histoire de la Sainte Russie,* 1854). This was the illustrated chronicle in which a startling hodgepodge of picturesque effects was arranged casually or with deliberate incongruity into a loose chronological sequence.

The most original contribution by the British was the creation by Charles Ross and his wife, Marie Duval, of "Ally Sloper," which appeared in *Judy* (1867–76). Ally was the first real continuing comic strip character; the schemes of this proletarian loafer were depicted in rather formless forerunners of the modern gag strip. He was then mock-gentrified by W.G. Baxter and W.F. Thomas in a maga-



Und schon ist er auf der Brücke,
Kracks! die Brücke bricht in Stücke;

Grad als dieses vorgekommen,
Kommt ein Gänsepaar geschwommen,
Welches Böck in Todeshast
Krampfhaft bei den Beinen faßt.



Wieder tönt es: „Meck, meck, meck!"
Plumps! da ist der Schneider weg!

Beide Gänse in der Hand,
Flattert er auf trocknes Land. —

"Max und Moritz" by Wilhelm Busch.

zine named after him, *Ally Sloper's Half-Holiday* (1884–1923). In *Fun*, another *Punch* imitator, J.F. Sullivan ran a series of attacks on working-class and petty bourgeois types under the title *British Working Man* (1878), and *British Tradesman* (1880).

Dominance of Wilhelm Busch

The dominant figure of the late 19th century is the German Wilhelm Busch, whose immense popularity in his own day has survived into the late 20th century. At first in periodicals, then in separately published albums, Busch quickly established himself as the first fully professional and truly popular comic-strip artist, appealing to educated and simple, young and old alike. Not being bound to journals, he could, like Töpffer, develop much longer stories than his French contemporaries, whose strips rarely exceeded 50 or so scenes running over three successive issues of a magazine. His graphic and narrative line appears more controlled, more predictable than that of Töpffer; it is comic in an earthier and more rational way. He revived the tradition of realistic social satire, although only two of his stories (*Helen* and *St. Anthony*) have a simply definable satiric butt (religious hypocrisy). Busch appears as a man obsessed with the farcical situation and its potential for physical violence. Happiness appears to lie in the avoidance of the petty annoyances of life and in the repression of instinctual behaviour. His cautionary tales of naughty children and animals may be regarded, on one level, as sophisticated parodies of the didactic juvenile literature of Germany and, on another, as condemnations of the childish sadism that is assumed to lie in everyone. On yet another level, all his work can be viewed as essays on the vulnerability of human dignity. His best known characters, the infant pranksters Max and Moritz, have spawned innumerable progeny down to our own day. Busch's graphic inventiveness was tremendous: his use of patterns of oscillation to represent movement and new conventional signs to express shock, pain, and other emotions constitute a vocabulary that has served the humorous strip cartoonist down to our own times. The rolling rhythm of Busch's graphic line has its counterpart in his facile comic verse, which is both independent and complementary. Variation in the amount of this verse accompanying each picture plays an essential part in the pace of the narrative.

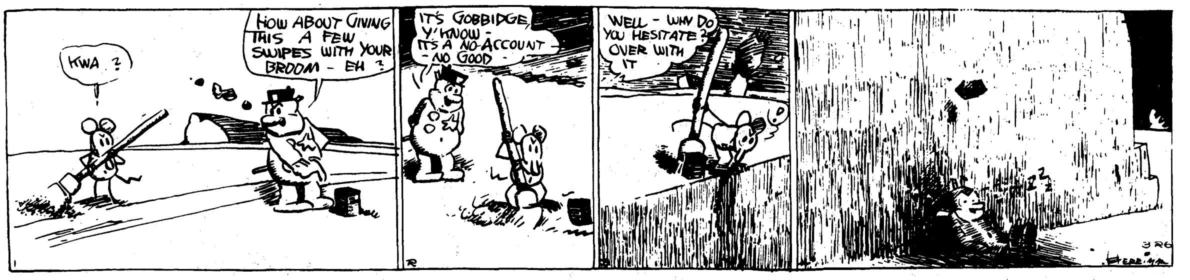
The only German follower of Busch worthy of the name was Adolf Oberländer, a sharp observer of human behaviour. The heir to Busch in France was Caran d'Ache, who in supplements to *Le Figaro* in the later 1880s drew the first strips to appear in a general-interest daily newspaper rather than in a weekly satirical magazine of relatively restricted readership. Always witty in a purely graphic sense, he frequently dispensed with captions altogether. In this respect, as in his technique of motif accumulation—his manner of letting a motif or movement snowball or crescendo ad absurdum—he taught much to later cartoonists, especially to Bateman in the 1920s.

Influence of the *Imagerie d'Épinal*

The influence of Busch on the development of the American newspaper strip was decisive; but European children's illustrated literature derived more from *Imagerie d'Épinal* and from the comic-strip artist Christophe. Throughout the 19th century the common people in France and The Netherlands, especially in rural areas, had subsisted on *Imagerie d'Épinal,* single broadsheets emanating from the town of Épinal, France, and hawked about the countryside. They covered, often in narrative form, such topics of folk interest as religious stories, patriotic histories, and fairy tales. The severe and simple didactic plates had a more or less realistic social emphasis. Some *Imagerie d'Épinal* are comic in content, although not always comic in style, relating, for instance, the folly of certain traditional social stereotypes or satirical characters from folklore and literature like Tyl Eulenspiegel and Baron Munchhausen. Christophe (pseudonym of Georges Colomb) raised this type of popular imagery to the level of the intelligent, urban child, first in the children's periodical, then in various albums published separately. These were originally designed, like Töpffer's, for the children of his own household and the pupils of his school. Christophe's gentle mockery of such types as the naïve bourgeois and the absent-minded professor is now implanted in French folklore. Christophe established a format for English and French children's comics which survived down to World War II, whereby the text is excluded from the image instead of being incorporated in the balloons of American usage.

First newspaper strips

20th century. *United States.* The modern newspaper strip was born in the heat of rivalry between giants of the American press. In January 1894 a comic strip filled, for the first time, a full-colour page of Joseph Pulitzer's newspaper the *New York World;* in October 1896 the publisher William Randolph Hearst announced in his rival paper, the *Morning Journal,* the first regular weekly full-colour comic supplement. This supplement ran to eight pages and included the "Yellow Kid" of Richard Outcault, whom Hearst had enticed away from the *New York World.* The "Yellow Kid" was the first continuous comic character in the United States. Outcault established earthy, strictly urban farce as the keynote of the early American strip, which thereafter grew in sophistication and sentimentality. The "Yellow Kid" also standardized the speech balloon, which had fallen largely into disuse since the 17th century and



"Krazy Kat," by George Herriman.

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its occasional appearance in the English caricatural strip around 1800. In 1897 Rudolph Dirks, at the instigation of Hearst, who had enjoyed the work of Busch as a child, worked up a strip based on "Max und Moritz," called the "Katzenjammer Kids," which proved an instant success. It had, for the first time, the fully developed form of the newspaper strip; *i.e.,* it used balloons, had a continuous cast of characters, and was divided into small regular panels (dispensing with the full panoramic scenes in which the "Yellow Kid" had appeared). The spread of comics to other newspapers was rapid and was aided by the development of newspaper syndication. The aesthetically outstanding strip of the early years was Winsor McCay's "Little Nemo in Slumberland" (first appearance 1906), which created a dreamworld at once gentle, exciting, and humorous. The strip was executed in fairy-tale illustration style, with a conscious display of colouristic effects.

The daily strip in black and white, indispensable to all major newspapers since 1915, was inaugurated in 1904 in the Chicago *American* with Claire Briggs' "A. Piker Clerk," in which a character gave tips on horse races. Harry C. "Bud" Fisher's "Mr. A. Mutt" (later "Mutt and Jeff") appeared in 1907. Initially a tip sheet, it soon became a general-interest comic.

Establishment of subject categories

During the years 1907–20 most of the major categories of American comics were established, including the first aviation, ethnic-character, and career-girl strips. The most important gag strip was George McManus' "Bringing up Father," dating from 1913. It was the first American strip to achieve international fame. Outstanding among the family-saga or domestic-problem strips that burgeoned during the '20s was Frank King's "Gasoline Alley," which dated from 1919. It strove for realism rather than farcical effects, and had a strict continuity (as opposed to the daily gag), during which, moreover, characters actually grew older. The first career-girl strip (Martin V. Branner's "Winnie Winkle") appeared in 1920. It was followed by the fashion-conscious "Tillie the Toiler" in 1921 by Russ Westover. Another major group of the '20s were fantastic, satirical, and parodistic. Elzie Crisler Segar's "Popeye" (first appearance 1919) still depended upon slapstick, but George Herriman's "Krazy Kat" (1911–44) created a tender world of poetry, at once surreal and humorous. Drawn with the greatest of graphic economy, the absurd interrelationships of a tiny cast of characters (basically three) were presented using the thinnest imaginable plot line. "Krazy Kat" was the first newspaper strip anywhere to be aimed at relatively intellectual adults.

During the 1930s the comics page expanded both in quantity of strips and in range of subject matter. Several of the strips created then have survived for more than 40 years. One of these, Chic Young's domestic comedy strip "Blondie" (first appearance 1930), has achieved unparalleled international renown.

A new category of immense significance emerged: the continuous-action adventure strip. This category took many forms: domestic and detective drama, science and space fiction, and, by 1938, war and supermen strips. The earliest adventure strip was "Tarzan" (first appearance 1929) whose creator Harold Foster broke completely with the prevailing caricatural style, adopted cinematic techniques, and sought picturesque, documentary realism. No less concerned with classic aesthetic effects was Alex Raymond, first master of the exotic space strip ("Flash Gordon," first appearance 1933). An aggressively cinematic, adventure strip, meticulously researched, was evolved by Milton Caniff in his "Terry and the Pirates," which appeared in 1934. Caricatural simplifications and grim forms of humour were introduced into the genre by Chester Gould's "Dick Tracy" (first appearance 1931), the detective strip par excellence, which is laced with science-fiction gadgetry and bizarre eroticism. Truly satirical forms of exaggeration returned to the strip with Capp's "Li'l Abner" (first appearance 1934).

The comic book

The demand for adventure stories spawned a new and highly lucrative vehicle for the comic strip: the cheap, staple-bound comic book. The first true comic books were marketed in 1933 as give-away advertising premiums. These had a 7½- by 10¼-inch (19- by 26-centimetre) page size, a format that has continued. By 1935 such titles as *Famous Funnies, Tip Top Comics, King Comics,* at first chiefly reprints of newspaper strips, then with original stories, were selling in large quantities. Specialization soon set in: *Detective Comics* (1937) and *Action Comics* (1938). "Superman," which appeared first in *Action Comics,* was the creation of Jerry Siegel (scenario or text) and Joe Shuster (art); it was soon syndicated and transposed to other media. The "Superman" formula of the hero who transcends all physical and social laws to punish the wicked, was widely imitated. The animated cartoon characters of Walt Disney also took root in the comic book.

World War II hastened the development of strips and comic books dealing with war and crime, the latter finding a new and avid readership among American soldiers stationed abroad. Being outside the control of newspaper editors, the comic book became increasingly violent and gruesome. The sadism of the American comic became proverbial; the "comic" became equated by Europeans with the "horror comic," and voices of educators were raised against it on both sides of the Atlantic. United States

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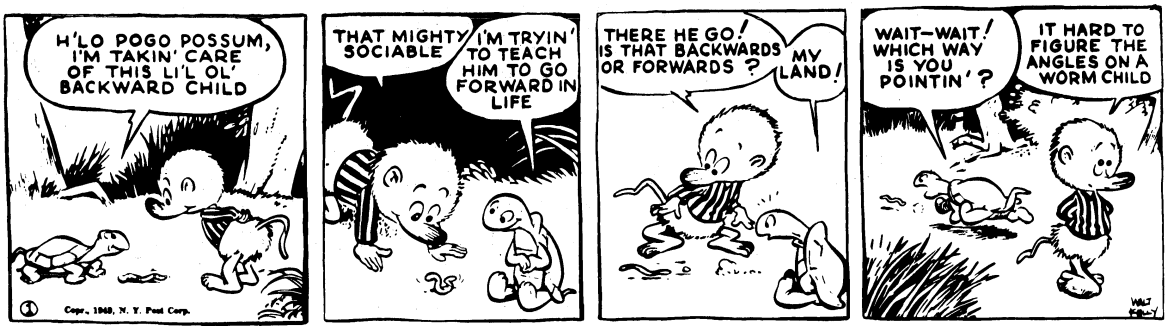
"Dick Tracy," by Chester Gould.



"Pogo," by Walt Kelly.
Reprinted courtesy Publishers—Hall Syndicate. © Walt Kelly

Congressional investigations in 1951 and 1954 blamed rising juvenile deliquency on the pernicious influence of the comic book. The industry responded by instituting systems of self-censorship, administered by several organizations; the more vicious-looking material was restrained, but, in Europe, some U.S. adventure strips continued to be criticized for their pursuit of violence and for their racist, militarist, and fascist tendencies.

Perhaps as a reaction, there was a parallel, postwar development in newspaper strips devoted to sentimental domestic drama—such as "Rex Morgan, M.D.," "Mary Worth," and "The Heart of Juliet Jones"—and simple-looking, but subtly conceived gag strips—such as "Beetle Bailey," the incorrigible draftee, and "Dennis the Menace," a sophisticated Katzenjammer. "Pogo" (first appearance 1946) by Walt Kelly was generally considered the most humorous and most cerebral. Drawn in a style reminiscent of Walt Disney, the strip transcended Disney's moral simplism, presenting a highly nuanced world populated with an almost Dickensian multitude of comic characters. "Pogo" exuded a tender, nostalgic air, perpetually ruffled by the breezes of sociopolitical allusion. Kelly's dialogue was also much admired, particularly since it appeared at a time when the comic strip was widely criticized for verbal poverty.

Literate strips of the 1950s

The literate strip with philosophical, psychological, and sociological overtones was the principal innovation of the later 1950s. Most successful in this group was Charles Schulz's "Peanuts" (first appearance 1950), a strip in which all the characters were children, dispensing with the adult foil. Of comparable psychological finesse, and imbued with truly satirical flashes, was Johnny Hart's "B.C." (first appearance 1958) and "Wizard of Id" in collaboration with Brant Parker (first appearance 1964), which had prehistoric and medieval settings, respectively. The major strip of political satire, "Feiffer" by Jules Feiffer (first appearance 1956), appeared weekly and was not run on the comics page but on editorial pages of the more liberal or "left" papers. In this strip the dialogue was more important than the drawings, which were deliberately repetitive; the content played upon the logic-twisting rhetoric of politicians and the neurotic relationships between social competitors and lovers.

The most significant innovation since about 1965 has been in the parodistic, erotic, and surrealistic comic. Parody and burlesque of the newspaper strip were pioneered by *Mad* magazine (first published in 1952). The erotic, laced with parody, found a home in the liberal periodical press ("Little Annie Fannie" in *Playboy* in 1962 and "Phoebe Zeitgeist" in *Evergreen Review* in 1968). Comic books satirizing a cultural underworld "liberated" by sex and drugs became popular, particularly among college students. These comics were pioneered in San Francisco by Robert Crumb, Rick Griffin, and S. Clay Wilson. The major drug-oriented strip, Gilbert Shelton's "Fabulous Furry Freak Brothers," circulated widely in the *Los Angeles Free Press*.

Europe. The first recurrent British comic characters, after Ally Sloper appeared in 1884, were Tom Brown's tramps "Weary Willie and Tired Tim." The strip was sponsored in 1896 by the publisher Alfred Harmsworth and was originally intended for the newly literate and semiliterate masses, but it developed into children's fare.

Strips for young children in Great Britain

Distinctive British contributions were the magazine of strips designed for preliterate children (*Tiger Tim's Weekly,* first published in 1920, and *Playbox,* which first appeared in 1925); the picture paper based on film comedy (*Film Fun, Kinema Comic*); and the multitude of children's magazines containing both articles and comic strips. The first strip for young children to appear in an adult newspaper was "Rupert, the Adventures of a Little Lost Teddy Bear," created by Mary Tourtel for the *Daily Express* in 1920. The text is fitted in below the balloonless pictures, in order to facilitate reading aloud by adults. The first British adult newspaper strip and after *Krazy Kat* the first daily strip anywhere designed exclusively for adults was the witty "Pop" by John Millar Watt, which first appeared in 1921. "Pop," together with Reginald Smythe's "Andy Capp" (first printed in 1957), were among the very few European strips to be exported to the United States. A notably original strip was Norman Pett's "Jane," published first in the *Daily Mirror* in 1932. It used an artful striptease theme and had great popularity with servicemen during World War II. The mildly satirical strip was pioneered in 1949 in the *Daily Express* by "Flook," a continuing narrative of various kinds of adventure (drawn by Wally Fawkes and written since 1958 by George Melly), and by Frank Dickens' "Bristow," a daily gag strip set in a business office. The outright satirical and political strip flowered in the new satirical magazines such as *Private Eye* (first published in 1961).

Comics in France and Italy

In France, Jean-Pierre Pinchon's "Bécassine" (first appearance 1905), modelled on Christophe, depicted the humorous adventures of a stupid but well-intentioned Breton servant; Louis Forton's "Bande des pieds-nickelés" (first appearance 1908), although ostensibly for children, had political touches and a mocking tone that appealed to adults as well. The first European strip to be fully developed in the American sense (notably as regards the use of balloons) was Alain de Saint-Ogan's "Zig et Puce" (first appearance 1925). France had no daily comic strip until 1934. There and in Italy, even more than in England, the market was smothered in the 1930s and '40s by American imports and imitations. An outstanding and hardy domestic product was "Tintin," created by the Belgian Hergé (pseudonym of Georges Rémi), a realistically conceived and relatively didactic adventure strip with a kind of Boy Scout hero. The immense popularity of "Tintin" has (since about 1964) been challenged by a warrior of ancient Gaul called "Astérix," the work of the writer (and comic-strip theorist) René Goscinny and the artist Albert Uderzo. "Astérix," besides being simply humorous and adventurous, indulges in sophisticated puns, witty anachronisms, and satirical flashes.

Colour, previously distorted and cheapened on the coarse newsprint of weekend papers and children's magazines, at last began to change the whole aesthetic character of the picture story—and improve it beyond recognition. In the 1960s Italian and French strips for adults by such artists as Guy Peellaert ("Jodelle"), Guido Crepax ("Valentina"), and Nicholas Devil ("Saga de Xam") paved the way for a new era in the comic strip, in which aesthetic expression aligned itself with contemporary art movements such as Pop, Op, and psychedelic art and availed itself of the most advanced printing techniques. The influence worked both ways. American art, particularly the Pop movement, was profoundly affected from about 1963 by comic-strip imagery. In Europe, where leading filmmakers admitted their debt to the comics, it was predicted that film, graphic art, and the comic strip would fuse into a new medium.

Dissemination of comics

The comics industry. The newspaper strip and comic book have become arguably the largest and most influential iconographic field in history, with literally millions of pictures produced since 1900. They certainly represent the dominant graphic mythology of the 20th century. Not even the film or television can boast of reaching a third of humanity, as can the comic strip. More than 100,000,000 Americans, young and old, educated and not, read one or more comic strips in their Sunday and daily newspapers. In 1963 there were more than 300 different strips in the United States. "Blondie" is syndicated in about 1,200 newspapers all over the world, "Peanuts" in about 1,000; "Pogo" reached and "Dick Tracy" reaches more than 50,000,000 readers in more than 500 newspapers. *Superman* comics circulated in the 1950s at the rate of 1,500,000 monthly; in 1943 U.S. comic books totalled 18,000,000 monthly copies, constituting a third of total magazine sales, to a value of $72,000,000.

Not surprisingly, reader participation has reached extraordinary heights; readers truly laugh and suffer with their favourite characters. Young received 400,000 suggestions for a name for Blondie's new baby.

Perhaps in no other form of art has the creator become to such an extent prisoner of his creation, to which he may be locked for his lifetime and which becomes in a real sense independent of his own existence, for the successful strip will almost always be continued by other artists if the original creator should die or lose interest in it. Traditionally, the newspaper strip is also in fief to the syndicates, publishers, and editors who regard it primarily as a circulation booster. The conventional view has been that it must not offend any conceivable readership or commercial-interest group and therefore must observe strict, conservative codes of morality and decorum. This explains the extent to which the newspaper strip generally has avoided controversial issues of the day. It was against such restrictions that the American underground comic and the European *bande dessinée pour adultes* ("adult strip") struck out, and in their wake some innovative newspaper strips, notably G.B. Trudeau's "Doonesbury," began to break new ground. (D.M.Ke.)

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Descartes and Cartesianism

René Descartes (Latin: Renatus Cartesius) is known as the father of modern philosophy. A 17th-century French mathematician, scientist, and philosopher, he was the first to oppose Scholastic Aristotelianism. He began by methodically doubting all forms of knowledge, then found certainty in the intuition that, when he is thinking, he exists; this he expressed in the famous statement "I think, therefore I am." He developed a dualistic system in which he distinguished radically between mind, the essence of which is thinking, and matter, the essence of which is extension in three dimensions. Descartes's metaphysical system is intuitionist, derived by reason from innate ideas, but his physics and physiology, based on sensory knowledge, are mechanistic and empiricist.

This article is divided into the following sections:

LIFE AND WORKS

Family and regional background. Descartes was born in La Haye (now Descartes), Fr., on March 31, 1596. Although La Haye is in Touraine, Descartes's family connections were south across the Creuse River in Poitou, where his father, Joachim, owned farms and houses in Châtellerault and Poitiers. Because Joachim was a councillor in the Parlement of Brittany in Rennes, Descartes inherited a low rank of nobility. Descartes's mother died when he was one year old. His father remarried in Rennes, leaving him in La Haye to be raised by his maternal grandmother and a nurse and probably also by his great-uncle Michel Ferrand, lieutenant general (court judge) in Châtellerault. The Descartes family was Roman Catholic, but Poitou was a Huguenot stronghold and Châtellerault a "secure city," in which the Edict of Nantes, which gave Protestants freedom of worship in France, was worked out in 1597–98. Descartes returned to Poitou regularly until 1628.

Jesuit education

Education, travels, and early influences. In 1606 Descartes was sent to the Jesuit college at La Flèche, established in 1604 by Henry IV. At La Flèche 1,200 young gentlemen were trained for careers in military engineering, the judiciary, and government administration. Besides classical studies, science, mathematics, and metaphysics, students were taught acting, music, poetry, dancing, riding, and fencing. Descartes's philosophy professor was Father François Véron, known later as the scourge of the Protestants. Aristotle was taught from Scholastic texts. In addition, Descartes received special attention from a relative, Father Charlet, later rector of La Flèche. In 1610 Descartes participated in an imposing ceremony, in which Henry IV's heart was placed in the cathedral of La Flèche. Henry IV's assassination had destroyed the hope of religious tolerance in France and Germany.

In 1614 Descartes went to Poitiers, where he took a law degree in 1616. At this time Huguenot Poitiers was in virtual revolt against Louis XIII. Descartes's father probably expected him to enter Parlement, but, because the legal age for that was 27, Descartes had seven years to wait. In 1618 he went to Breda in Holland for 15 months as a student in mathematics and military architecture in the peacetime army of the Protestant ruler, Maurice, prince of Orange. There Descartes met the physician Isaac Beeckman, who encouraged him in science and mathematics and for whom Descartes wrote his *Musicae Compendium* (written 1618, published posthumously; *Compendium of Music*).

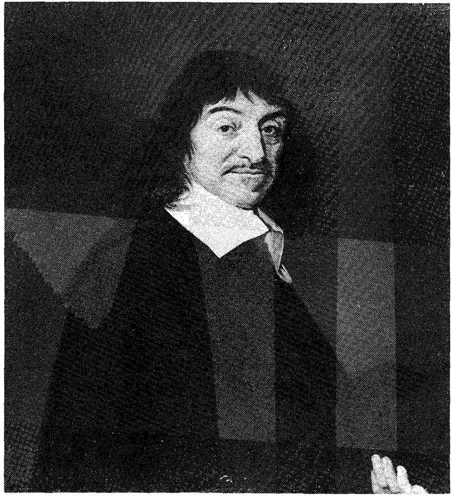
During the period 1619 to 1628 Descartes traveled in northern and southern Europe, saying that he was studying the book of the world. While in Bohemia in 1619 he had three dreams that defined for him his career as a scientist and a philosopher seeking knowledge for the benefit of humanity. By 1620 he had conceived of a universal method of deductive reasoning, applicable to all the sciences. He had also investigated reports of esoteric knowledge such as theosophical claims to command nature. Although disappointed with the followers of the magician Raymond Lulle and the alchemist Cornelius Agrippa, Descartes was impressed by the German mathematician and Rosicrucian Johann Faulhaber.

Affinities with Rosicrucian goals

Descartes shared a number of Rosicrucian goals and habits of life. Like Rosicrucians, he lived a single, secluded life, changing residence often (during his 22 years in Holland he lived in 18 different places), practiced medicine without charge, tried to increase human longevity, and expressed optimism about the ability of science to improve the human condition. At the end of his life he left a chest of personal papers—none of which have survived—with his close friend, the Rosicrucian physician Corneille van Hogelande, who handled his affairs in Holland. But Descartes rejected the Rosicrucians' magical and mystical beliefs. For him it was a time of hope for revolution in science. The English philosopher Francis Bacon, in *Advancement of Learning* (1605), had already proposed a new science of observation and experiment to replace the traditional Aristotelian science, as did Descartes later.

In 1620 Descartes was in the Roman Catholic army of Maximilian I, duke of Bavaria, who defeated the Protestants in Bohemia. There is, however, no evidence that Descartes ever participated in any battles. He said military life was idle, stupid, immoral, and cruel.

Cliche Musees Nationaux, Paris



Descartes, oil painting by Frans Hals, 1649. In the Louvre, Paris.

In 1622 Descartes moved to Paris. There he gambled, rode, fenced, and went to the court, concerts, and the theatre. Among his friends were the poets Jean-Louis Guez de Balzac, who dedicated his *Le Socrate chrétien* to Descartes in 1650, and Théophile de Viau, who was burned in effigy and imprisoned in 1623 for writing verses mocking religious themes. He also made friends with the mathematician Claude Mydorge and with Father Marin Mersenne, a man of universal learning who during his lifetime wrote thousands of letters to hundreds of scholars, writers, mathematicians, and scientists, keeping everyone aware—despite his almost unreadable handwriting—of what everyone else was doing. Mersenne was Descartes's main contact with the larger intellectual world. Descartes regularly hid from his friends in order to work, writing treatises, now lost, on fencing and metals. He acquired a high reputation long before he published anything.

Cardinal de Bérulle

At a talk in 1628 Descartes denied the alchemist Chandoux's claim that probabilities are as good as certainties in science and demonstrated his own method for attaining certainty. The Cardinal de Bérulle, who had founded the Oratorian teaching order in 1611 to rival the Jesuit order and who was forming the Compagnie du Saint-Sacrement, a militant, secret society of laymen to fight Protestantism, was impressed and invited Descartes to a conference. Bérulle was a strange combination of astute politician, courtier, and mystic who often advised the Queen Mother and talked familiarly with God and angels every day. Many commentators speculate that Bérulle urged Descartes to write an Augustinian metaphysics to replace Jesuit Thomism. There can be no question that in one way or another Bérulle tried to recruit Descartes to the Catholic cause. The result, however, was that within weeks Descartes left for Protestant Holland, took great precautions to conceal his whereabouts, and did not return to France for 16 years. Rather than taking Bérulle as director of his conscience, as some argue, it is probable that Descartes—who was a Roman Catholic but not an enthusiast, who was accused of being a Rosicrucian, who was from a Huguenot province, who glorified reason, and who advocated religious tolerance—was frightened by the mystical, militant Bérulle.

Choice of Holland as residence

Descartes said that he went to Holland to enjoy a greater liberty than was available anyplace else and to avoid the distractions of Paris and friends so that he could have the leisure and solitude to think. (He had inherited enough money and property to live independently.) Holland was a haven of tolerance. Descartes could be an original, independent thinker there without fear of being burned, for example, for giving natural explanations of miracles, as was Lucilio Vanini in 1619, or of being drafted as a soldier for the Roman Catholic Counter-Reformation. He opposed vows that restricted liberty and said, when accused of having illegitimate children, that, after all, he was a man and had taken no vows of chastity. In France, by contrast, religious intolerance was mounting. The Jews were expelled in 1615, and the last Protestant stronghold, La Rochelle, was crushed—with Bérulle's participation—only weeks before Descartes's departure. Catholic commentators insist that Descartes would have been safe in France, but the Parlement of Paris passed a decree in 1624 forbidding attacks on Aristotle on pain of death. Although the Catholic priests Mersenne and Pierre Gassendi did publish attacks without being persecuted, heretics continued to be burned, and laymen lacked church protection. Descartes may have felt in some jeopardy because of his friendship with such libertines as Father Claude Picot, a bon vivant known as the Atheist Priest, with whom Descartes left his financial affairs in France.

Residence in the Netherlands. In 1629 Descartes went to the university at Franeker, where he stayed with a Roman Catholic family and wrote the first draft of his *Meditations.* He registered at the University of Leiden in 1630, where he gained as a disciple the physician Henri Reneri. In 1631 he visited Denmark and in 1633–34 was in Germany with the physician and alchemist Étienne de Villebressieu, who invented siege engines, a portable bridge, and a two-wheeled stretcher. The physician Henri Regius taught Descartes's views at the University of Utrecht in 1639, starting a fierce controversy with the Calvinist theologian Gisbertus Voetius that continued until the end of Descartes's life. In his *Letter to Voetius* of 1648 Descartes made a plea for religious tolerance and the rights of man. He said that he wrote not only for Christians but also for Turks—meaning libertines, infidels, deists, and atheists. He argued that because Protestants and Catholics worship the same God, both can hope for heaven. But when the controversy became intense, Descartes sought the protection of the French ambassador and of his friend Constantijn Huygens, secretary to Prince Frederick Henry, ruler of the Dutch Republic.

In 1635 Descartes's daughter Francine was born to Helena Jans and was baptized in the Reformed Church in Deventer. Although Francine is referred to as Descartes's illegitimate daughter, her baptism is recorded in a register for legitimate births. Descartes said that his greatest sorrow was Francine's death of scarlet fever at the age of five and that he was not a philosopher who believed that one must refrain from tears to prove oneself a man.

The World, Rules,* and *Discourse on Method. In 1633 Descartes was about to publish *Le Monde* (published posthumously; *The World*), when he heard that the Italian astronomer Galileo Galilei had been condemned in Rome for publishing the view that the Earth revolves around the Sun. Because this Copernican position is central to Descartes's cosmology and physics, he suppressed *The World,* hoping that the church would retract its condemnation and make it possible for him to publish his work later. He feared the church, but he also hoped that his physics would one day replace Aristotle's in church doctrine.

Discourse on Method

In 1637 Descartes published *Discours de la méthode* (*Discourse on Method*), one of the first important modern philosophical works not written in Latin. Descartes said that he wrote in French so all who had good sense, including women, could read his work and learn to use their reason to think for themselves. He believed that everyone could tell true from false by the natural light of reason. In three essays forming part of the *Discourse* he illustrated his method for utilizing reason in the search for truth in the sciences. In *Optics* he then presented the law of refraction, in *Meteorology* he explained the rainbow, and in *Geometry* he gave an exposition of analytic geometry, which is a method of representing geometric figures with algebraic equations that made many previously unsolvable problems solvable. He also introduced the conventions of representing known numerical quantities with $a, b, c, \ldots,$ unknowns with $x, y, z, \ldots,$ and squares, cubes, and other geometric figures with numerals, as in $x^2, x^3, \ldots,$ which made algebraic notation much clearer than it had been before.

In *Discourse* and *Regulae ad Directionem Ingenii* (*Rules for the Direction of the Mind*), written by 1628 but not published until after his death, Descartes gave four rules for reasoning: (1) Accept nothing as true that is not self-evident. (2) Divide problems into their simplest parts. (3) Solve problems by proceeding from simple to complex. (4) Recheck the reasoning. These rules are a direct application of mathematical procedures. Descartes insisted that key notions and the limits of each problem must be clearly defined.

In *Discourse* he also provided a provisional moral code (later presented as final) for use while seeking truth: (1) Obey local customs and laws. (2) Make decisions on the best evidence and then stick to them firmly as though they were certain. (3) Change desires rather than the world. (4) Always seek truth. This code exhibits Descartes's prudential conservatism, decisiveness, Stoicism, and dedication. For Descartes all knowledge was like a tree—with metaphysics forming the roots, physics the trunk, and medicine, mechanics, and morals the branches—on which the fruit of knowledge is produced.

Meditations. In 1641 Descartes published in Latin—because it was dedicated to the Jesuit professors at the Sorbonne in Paris—*Meditationes de Prima Philosophia* (*Meditations on First Philosophy in Which Is Proved the Existence of God and the Immortality of the Soul*). Mersenne submitted it before publication to eminent thinkers, among whom were the Jansenist philosopher and theologian Antoine Arnauld, the English philosopher Thomas Hobbes, and the Epicurean atomist Gassendi. Mersenne collected their critical responses and published them under the title *Objectiones Septimae* (1642; *Objections and Replies*). Even though Descartes said that the Jesuit priest Pierre Bourdin, one of the respondents, was a fool, *Objections and Replies* is a landmark of cooperative discussion in philosophy and science at a time when dogmatism was the rule.

Cogito, ergo sum

Descartes begins *Meditations* with methodic doubt, rejecting as though false all forms of knowledge by which he was ever deceived. His arguments derive from the Pyrrhonism of the Greek skeptic Sextus Empiricus as reflected in the skeptical writings of Michel de Montaigne and Pierre Charron. Thus knowledge based on authority is set aside because even experts are sometimes wrong. Knowledge from sensory experience is declared untrustworthy because people sometimes mistake one thing for another, as with mirages. Knowledge based on reasoning is rejected as unreliable because one often makes mistakes as, for example, when adding. Finally, knowledge may be illusory because it comes from dreams or insanity or from a demon able to deceive men by making them think that they are experiencing the real world when they are not. Descartes finds certainty in the intuition that when he is thinking, even if deceived, he exists: *Cogito, ergo sum* ("I think, therefore I am"). The *cogito* is a logically self-evident truth that gives certain knowledge of a particular thing's existence: one's self. But the *cogito* justifies accepting as certain only the existence of the person who thinks it. If all one ever knew for certain was that he exists, and if one adhered to Descartes's method of doubting all that is uncertain, then he would be reduced to solipsism, the view that nothing exists but the individual and his thoughts. To escape this, Descartes argues that all ideas that are as clear and distinct as the *cogito* must be true, for if they were not, the *cogito* also, as a member of the class of clear and distinct ideas, could be doubted. But "I think, therefore I am" cannot be doubted. Therefore all clear and distinct ideas must be true.

On the basis of clear and distinct innate ideas Descartes then establishes that each mind is a spiritual substance and each body a part of one material substance. The mind or soul is immortal because it is unextended and cannot be broken into parts, as can extended bodies. Descartes also advances proof for the existence of God. He begins with the statement that he has an innate idea of God as a perfect being and then intuits that God necessarily exists, because if he did not, he would not be perfect. This ontological proof for the existence of God is at the heart of Descartes's rationalism, for it establishes certain knowledge about an existing thing solely on the basis of reasoning from innate ideas, with no help from sensory experience. Descartes then argues that because God is perfect, he does not deceive human beings; therefore the world exists. Thus Descartes claims to have given metaphysical foundations for the existence of his own mind, of God, and of the world.

Objections to Descartes's procedure

A famous objection to Descartes's procedure is Arnauld's Cartesian Circle, which exposes the circularity inherent in Descartes's reasoning. To know that God exists, one must trust the clear and distinct idea of God; but to know that clear and distinct ideas are true, one must know that God exists and does not deceive man. Descartes the rationalist rejected magic, but he failed to see that his ontological proof is word-magic based on the superstition that things can be determined by ideas and thoughts. In opposition to Descartes's rationalism, empiricists hold that descriptions of things must come after, not before, one knows by experience that they exist.

Physics, physiology, and morals. Descartes's goal was to be master of nature. He provided understanding of the trunk of the tree of knowledge in *The World, Optics, Meteorology,* and *Geometry* and revealed its roots in *Meditations;* he then spent the rest of his life working on the branches of mechanics, medicine, and morals. Mechanics is the basis of his medicine, or physiology, which in turn is the basis of his moral psychology. Descartes believed that all material bodies, including the human body, are machines that operate by mechanical principles. In his physiological studies he dissected animal bodies to show how their parts move. He argued that because animals have no souls, they do not think or feel; thus vivisection, which Descartes pioneered, is permitted. He also described the circulation of the blood but came to the erroneous conclusion that heat in the heart expands the blood, causing its expulsion. Descartes's *L'Homme, et un Traité de la formation du foetus* (*Man, and A Treatise on the Formation of the Foetus*) was published after his death.

In 1641 Descartes was visited by Picot and Jacques Vallée Desbarreaux, known as the Grand Debauché, who had published the libertine poet Théophile. Descartes used them as models for characters (he was himself model for a third) in his dialogue *Search After Truth.* In 1642 Samuel Sorbière, the French translator of Sextus and Hobbes, visited Descartes and wrote a charming description of him as host. Descartes then lived in the small but very elegant château of Endegeest, outside Leiden, near the court in The Hague.

Interaction of mind and body

In 1644 Descartes published *Principia Philosophiae* (*Principles of Philosophy*), a compilation of his physics and metaphysics. He dedicated this work to Princess Elizabeth, daughter of Elizabeth Stuart, titular queen of Bohemia, who was in exile in The Hague, for he had developed his moral philosophy in correspondence with her. According to Descartes, a human being is a union of mind and body, two dissimilar substances that interact in the pineal gland. He reasoned that the pineal gland must be the uniting point because it is the only non-double organ in the brain, and double reports, as from two eyes, must have one place to merge. He argued that each action on a person's sense organs causes subtle matter to move through tubular nerves to the pineal gland, causing it to vibrate distinctively. These vibrations give rise to emotions and passions and also cause the body to act. Bodily action is thus the final outcome of a reflex arc that begins with external stimuli and involves first an internal response, as, for example, when a soldier sees the enemy, feels fear, and flees. The mind cannot change bodily reactions directly—for example, it cannot will the body to fight—but it can change the pineal vibrations from those that cause fear and fleeing to those that cause courage and fighting.

Descartes furthermore argued that men can be conditioned by experience to have specific emotional responses. He, for example, had been conditioned to be attracted by cross-eyed women because he had loved a cross-eyed playmate as a child. But when he remembered this fact, he was able to rid himself of his passion. This insight was the basis for Descartes's defense of free will and of the mind's ability to control the body. But, despite such arguments in defense of free will, Descartes's *Les Passions de l'âme* (*Passions of the Soul*), dedicated in 1649 to Queen Christina of Sweden, holds that most of man's bodily actions are determined by external material causes.

Descartes's anti-Christian morality

Descartes's morality was anti-Christian in that, in contrast to Calvinists and Jansenists, he suggested that grace is not necessary for salvation but that human beings are virtuous and able to achieve salvation when they do their best to find and act upon truth. His optimism about the ability of human reason and will to find truth and reach salvation is in stark contrast with the pessimism of the Jansenist (predestinarian) apologist and mathematician Blaise Pascal, who believed that salvation comes only as a gift of God's grace. Descartes was accused of adhering to an error of the 12th-century Arabic philosopher Averroës, who said that virtuous behaviour depends on free will rather than on grace. Descartes agreed, however, that unless people believe in God and immortality they will see no reason to be moral.

Free will, Descartes stated, is the sign of God in human nature, and human beings can be praised or blamed according to their use of it. People are good only if they act in goodwill for the good of others; such generosity is the highest virtue. Descartes was Epicurean in his assertion that human passions are good in themselves and an extreme moral optimist in his belief that to understand the good is to want to do it; because passions are willings, to want something is to will it. But he was also a Stoic in his admonition that human beings should control their passions rather than change the world.

Although Descartes wrote no political philosophy, he approved of Seneca's admonition to acquiesce in the order of things. He rejected Machiavelli's recommendation to lie to friends, because friendship is sacred and life's greatest joy. Human beings cannot exist alone but must be parts of social groups, such as nations and families, and it is better to do good for the group than for oneself.

Descartes had been a puny child with a weak chest and was not expected to live. He therefore watched his health carefully and became a virtual vegetarian. In 1639 he bragged that he had not been sick for 19 years and expected to live to be 100. He told Elizabeth to think of life as a comedy; bad thoughts cause bad dreams and bodily disorders. Because there is always more good than evil in life, one can always be content, no matter how poorly off one is.

In his later years Descartes said that he had once hoped to learn to prolong life to a century or more. But he then saw that in order to achieve that goal the efforts of many generations would be required; he himself had not even learned to prevent a fever. Thus, he said, instead of continuing to hope for long life, he had found an easier way, namely to love life but not fear death. It is easy, he claimed, for a true philosopher to die tranquilly.

Final years and heritage. After 16 years in Holland Descartes returned to France for brief visits in 1644, 1647, and 1648, on financial business and to oversee the translation into French of *Principles, Meditations,* and *Objections and Replies* (the translators being, respectively, Picot, the Duke du Luyons, and Claude Clerselier). In 1647 he also met with Gassendi and Hobbes and suggested to Pascal the famous experiment of taking a barometer up Mount Puy-de-Dôme to determine the influence of the weight of the air. In Paris Descartes joined with Pierre d'Alibert, treasurer general of France, in a plan to establish a workshop school of arts and crafts in the Royal College. Picot returned with Descartes to Holland for the winter of 1647–48. During Descartes's final stay in Paris in 1648 the revolt of the nobility against the crown, known as the Fronde, broke out. As a result Descartes left Paris precipitously on Aug. 17, 1648, only days before his mortally ill old friend Mersenne died. Back at his retreat in Egmond, in Holland, Descartes was visited by the young Francis Burman, whose *Conversations* gives a genial and illuminating picture of Descartes.

Hector Pierre Chanut, Clerselier's brother-in-law, helped to procure a pension for Descartes from Louis XIV (which was never paid). Then Chanut, who was French resident and later ambassador to Sweden, gained an invitation for Descartes to the court of the Swedish monarch, Queen Christina, who by the close of the Thirty Years' War had become one of the most important and powerful monarchs in Europe. Descartes went reluctantly, arriving early in October 1649. He may have gone because he needed protection; the Fronde seemed to have destroyed his chances in Paris and the Calvinist theologians were still harassing him in Holland.

The 22-year-old Christina perversely made the 53-year-old Descartes rise at 5:00 AM to give her philosophy lessons, even though she knew of his habit of meditating in bed until 11 o'clock in the morning. She also ordered him to write verses for a ballet, *La Naissance de la paix* (1650; *The Birth of Peace*), a virtual pacifist tract celebrating Christina's role in the Peace of Westphalia, which ended the Thirty Years' War, and a comedy in five acts, now lost. In addition he wrote the statutes for a Swedish Academy of Arts and Sciences. While delivering these statutes to the Queen at 5:00 AM on Feb. 1, 1650, Descartes caught a chill. In this land, where he said that in winter men's thoughts freeze like the water, Descartes developed pneumonia. He died in Stockholm on Feb. 11, 1650. Many pious last words have been attributed to Descartes, but the most trustworthy report is probably that of his German valet, Schulter, who said that Descartes was in a coma and died without saying anything at all. The last thing Descartes wrote was a letter asking his brother to continue the pension Descartes had been paying to their old nurse.

Descartes's papers

After his death Descartes's papers came into the possession of Clerselier, a pious Catholic, who began the process of turning Descartes into a saint by cutting, adding to, and selectively publishing his letters. This cosmetic work culminated in 1691 in the massive biography by Father Adrien Baillet, who had previously published a 17-volume *Lives of the Saints.* But even while Descartes was still alive, there were questions as to whether he was a Roman Catholic apologist, primarily concerned with supporting Christian dogma, or an atheist, concerned only with protecting himself with pious sentiments while establishing a deterministic, mechanistic, and materialistic physics.

These questions remain difficult to answer, not least because many papers and manuscripts available to Clerselier and Baillet are now lost. The Roman Catholic Church made its decision in 1667 by putting Descartes's works on the Index of Forbidden Books on the very day his bones were ceremoniously placed in Ste. Geneviève-du-Mont in Paris. During his lifetime Protestant ministers in the Netherlands called him a Jesuit and a papist—*i.e.,* an atheist—but he said that they were intolerant, ignorant bigots. Up to about 1930 the majority of scholars, many of whom were religious men, believed that Descartes's major concerns were metaphysical and religious. By the late 20th century, however, numerous commentators had come to believe that Descartes was a Catholic in the way he was a Frenchman and a royalist, that is, by birth and by politics.

Descartes himself said that good sense is destroyed when one thinks too much of God. He once told the German protégée Maria van Schurmann that she was wasting her intellect studying Hebrew and theology. He also was perfectly aware of, although he tried to conceal, the atheistic potential of his materialist physics and physiology. Descartes also seemed indifferent to the emotional depths of religion. Whereas Pascal trembled when he looked into the infinite universe and perceived the puniness and misery of man, Descartes rejected the view that human beings are essentially miserable and sinful. Instead he exulted in the power of human reason to understand the cosmos and to promote human happiness. He held that it was impertinent to pray to God to change things, insisting rather that human beings must try to improve themselves.

CARTESIANISM

Certain versus probable knowledge

The Cartesian system. Cartesianism is a set of philosophical traditions and scientific attitudes. Metaphysically, Cartesianism is rationalist and Platonic, meaning that certain knowledge is derived by reason from innate ideas. This opposes the empiricist Aristotelian view that all knowledge is probable and is based on sense experience. In practice, however, Cartesians developed probabilistic scientific views from observation and experiment, as did empiricists. Cartesians had to be satisfied with uncertainty in science because they believed that God is omnipotent and that his will is entirely free. From this it follows that God, who, in addition to the material world, created all truths (such as those of mathematics and the laws of nature), could, nonetheless, given his infinite intellect and his free will, arbitrarily make even contradictions be true. The human intellect, by contrast, is finite; thus men can be certain only of the *cogito* and of revealed religion. Cartesians, however, did not derive scientific truths from religious knowledge, as did the Roman Catholic Church, and thus in practice they had to accept scientific knowledge as uncertain and probable.

Cartesians divide the world into a metaphysical dualism of two finite substances, mind (spirit or soul) and matter. The essence of mind is self-conscious thinking, the essence of matter extension in three dimensions. God is a third, infinite substance, whose essence is necessary existence, and God unites minds with bodies to create a fourth, compound substance, man. Humans have general knowledge of mind, matter, and God by contemplating innate ideas. But for particular knowledge of events in the world, humans depend on motions, transmitted from the sense organs through nerves to the brain, that cause sensible ideas to arise in the mind. Cartesians thus claimed to know the outer world by way of representative sensible ideas in the mind.

This dualism of mind and matter gives rise to serious problems concerning causal interaction and knowing. Given the essential unlikeness of mind and matter in the compound substance man, how can the body cause the mind to have sensible ideas? How can the mind cause the body to move? And how can the mind know the material world by way of sensible ideas that are mental? Various lines of Cartesian philosophy developed from different answers to these questions.

Cartesian mechanism. The first French Cartesians were physicists and physiologists who gave mechanistic explanations of physical and biologic phenomena. Father Nicolas Malebranche, a French theologian and philosopher who believed that animals are machines, once kicked a pregnant dog and then chastised such critics as Jean de La Fontaine, the French writer of animal fables, for expending their emotions over an unfeeling machine that moves and makes noises, depending merely on how and where it is stimulated, rather than in sympathy with human misery. In Paris the lectures of Pierre-Sylvain Régis on Cartesian physics—accompanied by spectacular demonstrations—created such a sensation that Louis XIV forbade them. Because Cartesianism challenged the traditional Aristotelian science supported by the Roman Catholic Church and because the church stood behind the divine right of kings to rule, the king feared that any criticism of traditional authority might give rise to revolution. (Descartes's stress on each individual's ability to think for himself did provide support for republicanism in the 18th century.)

Advancement in mechanical arts and crafts provided the practical foundation of Cartesian mechanism. In the 17th century mechanical inventions such as statues that walked and talked by application of levers and pullies and organs that played by waterpower were well known. Pascal invented a calculating machine based on principles worked out by clockmakers and by inventors of spinning and knitting machines, such as the Englishman William Lee. The first Cartesian inventors were the French craftsman Ferrier, who made lenses according to Descartes's designs, and Étienne de Villebressieu, who with Descartes's collaboration developed an improved pump.

Mechanism was promoted by Mersenne and derived theoretically from Gassendi's Epicurean atomism and Galileo's experiments with moving bodies. According to Descartes the material universe consists of an indefinitely large plenum of infinitely divisible matter separated into the subtle matter of space and the denser matter of bodies by a set quantity of motion imparted by God. Bodies swirl like leaves in a whirlwind in vortices as great as that in which the planets sweep around the Sun and as small as that of tiny, spinning globes of light. All bodily joinings and separations are mechanical and result from moving bodies bumping into one another. Because the amount of motion is conserved according to the laws of nature, the Cartesian material world is deterministic. After the initial impulse the world evolves lawfully; if the speeds and amounts of motion and the positions of all the whirling portions of matter in the universe were described for any one time, then simple deductions with reference to the laws of motion would allow their descriptions for any other time. But of course only God has the infinite intellect required to make these calculations.

Cartesian determinism

God is the primary cause of the existence of the material universe and of the laws of nature. But all material events—*i.e.,* the actual movements and interactions of bodies—occur as results of secondary causes. God stands merely for the uniformity and consistency of the laws of nature.

Cartesianism was popularized in England by Antoine Le Grand, a French Franciscan, who gave an exposition of the typically ingenious mechanistic explanation of light and colour. Light consists of tiny, spinning globes of highly elastic, subtle matter that fly through the air in straight lines and bounce like tennis balls on angles consistent with the laws of optics. Different colours, in this view, are caused by different speeds and spins of the globes, determined by the texture of the surfaces that reflect or transmit light. The spectrum of colours caused by light passing through a triangular prism is the result of the globes passing more slowly through thicker rather than thinner portions. The same spectrum of colours occurs when light passes through thinner and thicker parts of raindrops, giving rise to rainbows. Such simple mechanistic explanations were shown by Newton and others to be inadequate for explaining the forces of gravitation and chemical bonding. Nonetheless, this explanation of light and colour is in principle like the explanation, now accepted, that light can be separated into colours according to different wavelengths.

The challenge of Newton's physics

By the end of the 17th century Cartesian rationalistic physics had been abandoned for the empirical results of observation and experience, and the explanatory value of Newton's mathematical physics vastly exceeded that of simple mechanism. Cartesians admitted that Descartes's laws of motion were wrong and that the principle of the conservation of motion had to be replaced by the Newtonian principle of the conservation of energy. Although Jacques Rohault's *Treatise* (1671) on Cartesian physics was translated into English in 1723 by Samuel Clarke, a philosopher and theologian and also a friend and disciple of Newton's, and by his brother Thomas Clarke, their corrections and annotations with footnotes turned the treatise into an exposition of Newtonian physics.

Mechanism versus Aristotelianism. Cartesian mechanism was in opposition to the Scholastic Aristotelian science supported by both Roman Catholic and Protestant theologians. These theologians held that because all things are created by God with a given nature, there is no evolutionary development of either the universe or animals. For Aristotelians the motive force of a thing is the internal power of its nature—on the model of willpower—that, for example, makes an acorn become an oak tree. Every individual thing was thus thought to contain an essence or nature, a spirit or soul, that gives it its distinctive being. These souls are what give life to living things, whereas for mechanists, animate bodies are no more than very complicated machines. For Cartesians, the human body, for example, does not die because the soul leaves it, as the Aristotelians say; the soul leaves because the body disintegrates. According to the Aristotelian view, all things have virtues or desires and strive to reach predetermined goals. Thus everything is purposive, and to understand a thing scientifically one must know the end or final cause toward which it aims.

Rejection of the occult

Descartes rejected both the teleological, animistic view and the related theory of alchemy that there are vital forces in things. Cartesians denied the existence of these occult, or magic, forces, insisting instead that only God and humans have spirits, wills, purposes, and ends. They perceived both animate and inanimate bodies as having no goals but as simply being pushed around passively. For Cartesians, science therefore consisted not of looking for final causes but rather for the mechanical laws of moving bodies. Thus Descartes ridiculed the belief that the stars—bodies many times larger than the Earth and immensely far away—exert influences on human beings, as astrologers claim. (Descartes's attack on astrology was made at a time when the Cardinal de Richelieu, Louis XIII's first minister, had horoscopes cast and consulted seers such as Bérulle.) Mechanists also opposed magic by conjuring or magic by incantations—that is, appeals to angels and the supposed spirits in things—and sympathetic magic, because they perceived no mechanical reason why things that look alike should influence one another.

There are no mysteries in Cartesian mechanistic natural science. By insisting on human free will, however, Descartes places the human soul or mind, like God, outside deterministic nature. Because the body is a part of deterministic nature, the mind's control of body movements tends to be miraculous and destructive of mechanistic determinism. In the end, given Descartes's system, the human mind must be perceived as an animistic force and its ability to move the body by willpower as magic.

Mind, body, and man. Most Cartesians believed that the mind and body interact. When asked how, Régis gave the standard Cartesian reply that human beings experience it and that God can and does make interaction take place, even if one cannot understand how. This is to offer a nonphilosophical, mystical answer to a problem that demands a clear explanation. As for the question of how ideas represent objects, the physician Rohault spoke for all Cartesians when he replied that God made ideas that are mental to represent material bodies without resembling them and that no further explanation was necessary. Again, this answer abdicates philosophy for mysticism and theology; it sounds like the reintroduction of Aristotelian forms or natures and thus goes against Descartes's own claim to be able to give intelligible explanations of all natural phenomena.

Thomists (adherents of the philosophical and theological system developed by Thomas Aquinas) say the soul or mind is an Aristotelian form or nature, *i.e.,* a force or power to inform or make neutral matter into a specific thing. In the example of the acorn, the form is what makes the acorn develop into an oak tree. On this view, a human is a unitary substance compounded of a form (the soul) and neutral matter, neither of which can exist independently. The human soul, however, is said to be a substantial form, miraculously able to exist independently of matter and thus to survive the death of the body. Descartes pointed out that this notion of substantial form is contradictory. Cartesian dualism presents the mind or soul as a substance existing in itself, independently of matter, and thus explains immortality without resorting to the miracle of turning the soul-form into a substance. This dualism, however, presented Cartesians with a serious problem about the ultimate nature of humans. The French physician Louis de La Forge concluded that at death the mind or soul is completely severed from all knowledge of individual bodies. This occurs because sensible ideas, for Cartesians, arise from the mind–body union for the sole purpose of preserving the body by presenting harmful things as painful and beneficial things as pleasurable. Human beings learn by experience what to seek and avoid, and the memory of these experiences is preserved in the brain. But once the body dies, both the need for sensible ideas and their memory traces in the brain are destroyed.

The soul after death

What the soul knows of matter after death is only the general idea of extension; that is, it knows mathematics but cannot remember the faces of friends. One problem this view raises is that, because bodily associations and memories are eliminated at death, individual personality is lost; thus it would be impossible to differentiate one soul from another. On Cartesian principles, a human being survives death only as an impersonal soul, identical to all other bodiless souls. Like the notion that animals are mere machines, the Cartesian conclusion that sensible manifestations of this life are neither continued nor remembered in the next was unpopular.

Science and religion. Besides the mind–matter dualism within Cartesian metaphysics, in Cartesianism as a whole there is a dualism between rationalist metaphysics, which depends on the certainty of general reasoning about innate ideas of mind, matter, and the necessary existence of God, and mechanistic physics, which advances scientific knowledge by accumulating probabilities based on observation and experience of the particularities of the material world. This has led some commentators to present Descartes almost exclusively as an apologist for Christianity, while others have argued persuasively that he was an atheist materialist interested only in physics. Descartes publicly denied interest in theology, but in letters he offered mechanistic explanations of transubstantiation. Thomists said that the forms of bread and wine are miraculously sustained as substantial forms (like the Thomistic soul), while their matter is replaced by Christ's flesh and blood. Rohault utilized the Cartesian view that sensible ideas are caused by configurations of the parts of material bodies to argue that if bread and wine were replaced by flesh and blood whose parts have exactly the same configurations as bread and wine, the flesh and blood would look, feel, and taste like bread and wine. While this would require miraculous replacement of bread and wine by flesh and blood, there would be no issue of contradictory substantial forms.

Arnold Geulincx's deterministic Cartesian ethics

The Flemish Calvinist Arnold Geulincx developed a deterministic Cartesian ethics. In his view, although one can do only what God has willed, one is free to accept this willingly or unwillingly. Virtue consists in the humble, diligent, and obedient acceptance of the justice of God's decrees in the light of reason, whereas sin and evil result from man's egotistic (and futile) stand against God. This Stoical ethics, with its affinity to Calvinist and Jansenist predestinarianism, is as deterministic as Cartesian physics; it does, however, contradict Descartes's claim that the human will is free not just to accept or reject the rightness of predetermined bodily actions but also to choose and cause a particular action.

Malebranche and occasionalism. The most important philosophical work stemming directly from Descartes's writings is *Search After Truth* (1674–75) by the Oratorian Malebranche. His position is known as occasionalism, and it was adopted also by Geulincx and the French philosopher Géraud de Cordemoy. Malebranche was convinced by the argument—urged most strongly by the French skeptic and chaplain Simon Foucher—that Cartesian mind and matter because of their essential unlikeness cannot interact. Malebranche held that on every occasion when human bodies interact with the world, God provides the appropriate sensible ideas in their minds. And on every occasion when humans will that their bodies move, God makes them move. Thus there is no direct causal interaction between mind and body; there are only separate but parallel sequences of mental and material events intermediated by God. Foucher also argued that, because ideas that are mental cannot resemble material things, ideas cannot represent bodies. That is, sensible ideas are coloured images, tactile feelings, sounds, odours, and tastes—as sensed—while the properties of material bodies are size, shape, position, and motion or rest; for example, the taste of sweetness is utterly unlike the sugar that causes it. How then do these ideas give man knowledge of the outer world as it really is? Like Descartes before him, Malebranche denied that ideas must resemble their objects to represent them. To the possibility that one might have sensible ideas of a nonexistent world, Malebranche said tersely that the first chapter of Genesis assures the existence of the material world. As to how man's ideas of this world are true, Malebranche offered the Platonic view that ideas of all things reside in the mind of God and that on appropriate occasions God illuminates these ideas for human observation. Thus human beings see all things in God and rest assured in his goodness.

Malebranche denied that he was a Cartesian. Unlike Descartes, he argued that introspection gives no knowledge of the essence of the mind. This led the English philosopher John Locke to suggest that matter might, for all human beings know, be able to think. All Cartesians opposed this possibility because the separation of mind or soul from body supports the Christian doctrine that the human soul survives the death of the body.

Later philosophers. The rationalist metaphysics of the Dutch-Jewish philosopher Benedict de Spinoza derive from Descartes. Spinoza's *Ethics* (1677) was written in mathematico-deductive form, with definitions, axioms, and derived theorems. Spinoza's metaphysics is monistic, pantheistic, and deist in that he argued that there is only one substance, God, which is the world and not (as theists hold) a person. This substance has an infinite number of attributes, each of which expresses the totality of God or the world. The only attributes humans know are mind and matter. All attributes are parallel in every respect, and thus, although mind and matter do not interact, for Spinoza as for Malebranche they appear to do so.

Another rationalist, the German scientist and diplomat Gottfried Wilhelm Leibniz, also gave a parallelistic answer to mind–body problems.

Leibnizian monads

Leibnizian monads, or psychic units of reality—inner forces, or powers, of the sort Descartes banished from the world—are self-complete. Each monad reflects the entire universe from a distinctive point of view. Monads do not interact, but because each is in preestablished harmony with all the others, the appearance of interaction is maintained.

The Irish phenomenalist George Berkeley also presented a monistic system. He vindicated the common-sense belief that material things are like sensible ideas by saying that there is no matter; bodies are only collections of sensible ideas provided to the human mind in lawful order by God. Thus ideas are not representative and open to skeptical objections about whether they provide true knowledge of things. Things or phenomena are known directly, according to Berkeley. There also is no problem of mind–body interaction because bodies are made up of mental ideas, some of which the human mind can control.

Hobbes, by contrast, did away with mind, asserting that only matter exists. Thinking and sensing mind is just a manifestation of brain activity. Thoughts are epiphenomenal (*i.e.,* above and irrelevant to mechanical material interactions); they are like the noise of a squeaky wheel. Thoughts and feelings reflect what is going on in the body just as the squeak indicates that the wheel needs greasing. But thoughts are totally derivative and have no influence whatsoever on the operation of the body machine. Thus Hobbes's monistic materialism also does away with the problem of mind–body interaction.

The way of ideas and the self. Two further lines of influence can be traced to Descartes, one stemming from the Cartesian theory of knowing by way of representative ideas, the other from the Cartesian *cogito.* The former is rooted in Galileo's distinction between real, or primary, properties of bodies—size, shape, position, and motion or rest, all of which are quantitative—and sensible, or secondary, properties—colours, tactile feelings, sounds, odours, and tastes—said to exist only in the mind. As Descartes's theory of light shows, and as Locke later specified, secondary properties of bodies are only distinctive arrangements of primary properties that act on sense organs, causing sensible ideas in the mind. Locke thought that the sensible ideas of size, shape, position, and motion or rest resemble material objects as they are. (Modern science is based almost exclusively on mathematical descriptions of such primary properties.) But one might again object that, because all ideas are mental, none can resemble material objects. How does one know that ideas represent bodies? As stated above, Berkeley's phenomenalism—that only minds and sensible ideas exist—is one heroic solution to this skeptical problem: Bodies are known directly because they are made of sensible ideas. The Scottish philosopher David Hume took this solution one step further by saying that minds also are nothing but collections of ideas. Building on Hume in the 20th century, the English philosophers G.E. Moore and Bertrand Russell, the German positivists Moritz Schlick and Rudolf Carnap, and the Austrian-born Ludwig Wittgenstein called sensible ideas sense data, out of which they made logical constructions of the world. The German philosopher Edmund Husserl attempted to establish a science of phenomenology by describing sensible ideas. Russell and the American pragmatist William James suggested constructing both mind and matter out of neutral monads. All of these systems stem from the Cartesian way of ideas.

The second line of influence derives from the Cartesian *cogito,* with its roots in the Neoplatonic philosophy of St. Augustine. Stress on the self or ego is behind the developmental idealism of the German philosopher G.W.F. Hegel, who envisioned a World Soul coming to consciousness. Focus on the being of the self by the German philosopher Martin Heidegger led to the existentialism of the French philosopher Jean-Paul Sartre, who argued that each individual chooses to come into being out of nothing. Sartre also upheld the Cartesian position that the self is conscious by denying the unconscious proposed by the Austrian psychoanalyst Sigmund Freud.

Contemporary influences. Some aspects of Cartesian metaphysics are still strongly defended in the 20th century. Like Descartes, the American linguist Noam Chomsky argues that all human beings have an innate ability to learn language which distinguishes humans from all other animals. The Australian physiologist John C. Eccles and the English primatologist Wilfred E. Le Gros Clark postulate the mind as a nonmaterial entity. The British philosopher Karl Popper proposes a dualism between the material and the ideational.

20th-century attacks on Cartesian dualism

The strongest 20th-century attack on Cartesian dualism was launched by the British analytic philosopher Gilbert Ryle in *The Concept of Mind* (1949), where he exposes what he describes as the fallacy of the ghost in the machine. He argues that the mind—the ghost—is simply the intelligent behaviour of the body. Like many contemporary analytic philosophers, Ryle maintains that metaphysical questions about being and reality are nonsense because they include reference to empirically unverifiable entities. His position, like that of the Australian philosopher J.J.C. Smart, is ultimately materialist: The mind is the brain. The American pragmatist Richard Rorty in *Philosophy and the Mirror of Nature* (1979) argues that the Cartesian demand for certain knowledge is a holdover from man's mistaken quest for God. Rorty says that philosophy in the Cartesian tradition is the 20th century's substitute for theology and should, like God, be gently laid to rest.

But Descartes's influence is so pervasive that all Western philosophers, even when they deny Cartesianism, can be said to be Cartesians, as they can be said to be Greeks; their positions are necessarily responses to issues posed by Descartes. Descartes also stands at the beginning of modern mathematics; through his invention of analytic geometry he laid essential ground for the invention of the infinitesimal calculus by Newton and Leibniz. Cartesian method was brilliantly elaborated by the Jansenists Pierre Nicole and Arnauld in logic and grammar texts that are fundamental to linguistics. Hume's distinction between fact and value issued from the elevation of the mathematically objective over the emotional and subjective. Descartes's skeptical, mathematical method underpins modern science; his rationality forms Western man's modern consciousness; and his intense desire to control mind and matter corresponds to the ultimate secular goal of contemporary science and society. This stress on mastery of nature, including man, has led to the contemporary sense of "Cartesian" as standing for everything that is crassly materialistic, logical, unfeeling, and inhuman in science, technology, and society. But Descartes himself said that he wanted only to serve humanity.

MAJOR WORKS

The history of the original works and their early translations into English is as follows: *Musicae Compendium* (written 1618, published posthumously; *Renatus Des-Cartes Excellent Compendium of Musick,* 1653); *Regulae ad Directionem Ingenii* (written 1628, published posthumously); *Le Monde de Mr Descartes, ou Le traité de la lumière* (written 1633, published posthumously); *Discours de la méthode pour bien conduire sa raison, & chercher la verité dans les sciences. Plus la dioptrique; les meteores; et la geometrie* (1637; *A Discourse of a Method for the Wel-guiding of Reason, and the Discovery of Truth in Sciences,* 1649); *Meditationes de Prima Philosophia* (1641; and its 2nd ed., with *Objectiones Septimae,* 1642; *Six Metaphysical Meditations; Wherein It Is Proved That There Is a God,* 1680); *Principia Philosophiae* (1644); *Les Passions de l'âme* (1649; *The Passions of the Soule,* 1650).

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(R.A.W.)

Domestic Cats

All cats are grouped in the family Felidae. From the largest, the tiger, to the smallest, the house cat, felids are characterized by supple, low-slung bodies, finely molded heads, long tails that aid in balance, and specialized teeth and claws that adapt them admirably to a life of active hunting. The family of cats belongs to the mammalian order Carnivora, the "flesh eaters."

Domestic cats (*Felis catus*) possess the features of their wild relatives in being basically carnivorous, remarkably agile and powerful, and finely coordinated in movement. It is noteworthy that the ancestors of the other common household pet, the dog, were social animals that lived together in packs in which there was subordination to a leader, and the dog has readily transferred its allegiance from pack leader to human master. The cat, however, has not yielded as readily to subjugation. Consequently, the house cat is able to revert to complete self-reliance more quickly and more successfully than most domesticated dogs.

For an account of the relationship of the family of cats to other carnivores, see MAMMALS: *Carnivora.*

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 313, and the *Index.*

The article is divided into the following sections:

ORIGIN AND HISTORY OF CATS

The "cat pattern," established very early in the evolution of modern mammals, was a successful one: early cats were already typical in form at a time when the ancestors of most other modern mammalian types were scarcely recognizable. They first appeared in the Early Pliocene epoch (5.3 to 3.4 million years ago), and they have continued with remarkably little change into modern times.

Domestication. Although its origin is hidden in antiquity, the domestic cat has a history that dates nearly 3,500 years to ancient Egypt. There are no authentic records of domestication earlier than 1500 BC, but it may have taken place sooner. Although the cat was proclaimed a sacred animal in the 5th and 6th dynasties (*c.* 2465–*c.* 2150 BC), it had not necessarily been domesticated at that time. It is probable that the Egyptians domesticated the cat because they realized its value in protecting granaries from rodents. Their affection and respect for this predator led to the development of religious cat cults and temple worship of cats.

Cats have long been known to other cultures. Wall tiles in Crete dating from 1600 BC depict hunting cats. Evidence from art and literature indicates that the domestic cat was present in Greece from the 5th century BC and in China from 500 BC. In India cats were mentioned in Sanskrit writings around 100 BC, while the Arabs and the Japanese were not introduced to the cat until about AD 600. The earliest record of cats in Britain dates to about AD 936 when Howel Dda, prince of south-central Wales, enacted laws for their protection.

Ancestry of breeds

Even though all cats are similar in appearance, it is difficult to trace the ancestry of individual breeds. Since tabbylike markings appear in the drawings and mummies of ancient Egyptian cats, present-day tabbies may be descendants of the sacred cats of Egypt. The Abyssinian also resembles pictures and statues of Egyptian cats. The Persian, whose colouring is often the same as that of mixed breeds (although the length of hair and the body conformation are distinctive), was probably crossed at various times with other breeds; the tailless Manx cat, like the hairless Sphynx cat and curly-coated Devon rex, is a mutation. The ancestry of Persian and Siamese cats may well be distinct from other domestic breeds, representing a domestication of an Asian wild cat (the ancestor of the Egyptian cat is believed to have come from Africa). In fact, nothing is known of the ancestry of the Siamese types, and there is no living species of Asian cat that would serve as ancestor.

Associations with human culture. The cat has long played a role in religion and witchcraft. In the Bible, "cat" is mentioned only in the apocryphal Letter of Jeremiah. The cat figured prominently in the religions of Egypt, the Norse countries, and various parts of Asia. The Egyptians had a cat-headed goddess named Bast. Thousands of cat mummies have been discovered in Egypt, and there were even mouse mummies, presumably to provide food for the cats. Often the cat has been associated with sorcery and witchcraft, and the superstitions regarding cats are innumerable. Throughout the ages, cats have been more cruelly mistreated than perhaps any other animal. Black cats in particular have long been regarded as having occult powers and as being the familiars of witches.

The cat is a familiar figure in nursery rhymes, stories, and proverbs. The English legend of Dick Whittington and his cat is a particular favourite. The writers Théophile Gautier and Charles Baudelaire paid it homage, and in the 20th century Rudyard Kipling, Colette, and T.S. Eliot wrote of cats.

GENERAL FEATURES AND SPECIAL ADAPTATIONS

The average weight of the household cat varies from 6 to 10 pounds, although among nonpedigreed cats weights up to 28 pounds are not uncommon. Average lengths are 28 inches for males and 20 inches for females. In keeping with a carnivorous habit, the cat has a simple gut; the small intestine is only about three times the length of the body.

The skin of the cat, composed of dermis and epidermis, regenerates and fights off infection quickly. Tiny erector muscles, attached to hair follicles, enable the cat to bristle all over. Thus, although the cat is a relatively small animal, it can frighten enemies by arching its back, bristling, and hissing.

Coordination and musculature. Cats are among the most highly specialized of the flesh-eating mammals. Their brains are large and well-developed. Cats are digitigrade; that is, they walk on their toes. Unlike the dog and horse, the cat walks or runs by moving first the front and back legs on one side, then the front and back legs on the other side; only the camel and the giraffe move in a similar way. The cat's body has great elasticity. Because the vertebrae of the spinal column are held together by muscles rather than by ligaments, as in humans, the cat can elongate or contract its back, curve it upward, or oscillate it along the vertebral line. The construction of the shoulder joints permits the cat to turn its foreleg in almost any direction. Cats are powerfully built animals and are so well coordinated that they almost invariably land on their feet if they fall or are dropped.

Teeth. The cat's teeth are adapted to three functions: stabbing (canines), anchoring (canines), and cutting (molars). Cats have no flat-crowned crushing teeth and therefore cannot chew their food; instead, they cut it up. Except for the canines and molars, the cat's teeth are more or less nonfunctional; most of the cheek teeth do not even meet when the mouth is closed. The dental formula in all cats, for either side of both upper and lower jaws, is incisors 3/3, canines 1/1, premolars 3/2, and molars 1/1. The total number of teeth is 16 in the upper jaw and 14 in the lower. Primary, or milk, teeth number 24; these are replaced by the permanent teeth at about five months. Each half of the jaw is hinged to the skull by a transverse roller that fits tightly into a trough on the underside of the skull, making grinding movements impossible even if the cat had teeth suitable for grinding.

Retraction of claws

Claws. There is a remarkable mechanism for retracting the cat's claws when they are not in use. The claw is retracted or extended by pivoting the end bone of the toe, which bears the claw, over the tip of the next bone. The action that unsheathes the claws also spreads the toes widely, making the foot more than twice as broad as it normally is and converting it into a truly formidable weapon. This claw-sheathing mechanism is present in all species of the cat family except the cheetah. Although there are no nerve endings in the nail itself, blood capillaries are present in the inner part.

Senses. Cats are generally nocturnal in habit. The retina of the cat's eye is made extra sensitive to light by a layer of guanine, which causes the eye to shine at night in a strong light. The eyes themselves, large with pupils that expand or contract to mere slits according to the density of light, do not distinguish colours clearly. Cats have a third eyelid, or nictitating membrane, commonly called the haw. Its appearance is used frequently as an indicator of the cat's general state of health.

The cat's sense of smell, particularly well developed in the adult, is crucial to its evaluation of food, so that a cat whose nasal passages become clogged as a result of illness may appear to lose its appetite completely. Cats can distinguish the odour of nitrogenous substances (*e.g.,* fish) especially keenly.

The sense of touch is acute in cats. The eyebrows, whiskers, hairs of the cheek, and fine tufts of hair on the ears are all extremely sensitive to vibratory stimulation. The functions of the whiskers (vibrissae) are only partially understood; however, it is known that, if they are cut off, the cat is temporarily incapacitated. The toes and paws, as well as the tip of the nose, are also very sensitive to touch.

Cats also have an acute sense of hearing. Their ears contain almost 30 muscles (compared with 6 in humans); as a result, they can turn them many times more quickly in the direction of a sound than can a dog. The ears of cats, although receptive to ultrasonic frequencies up to 25,000 vibrations per second, are slightly inferior to those of dogs, which register 35,000 vibrations per second.

BEHAVIOUR

Special traits. The cat has a subtle repertoire of facial expressions, vocal sounds, and tail and body postures that express its emotional state and intentions. These various signals serve to increase, decrease, or maintain social distance. One distinctive social behaviour involves rubbing the side of the head, lips, chin, or tail against the owner and against furniture. These regions of the cat's body contain scent glands that seem to play a role in establishing a familiar odour in the cat's environment.

The tongue of all cats, which has a patch of sharp, backward-directed spines near the tip, has the appearance and feel of a coarse file; the spines help the cat to lap up liquids and also to groom itself. The disposition to cleanliness is well established in cats, and they groom themselves at length, especially after meals.

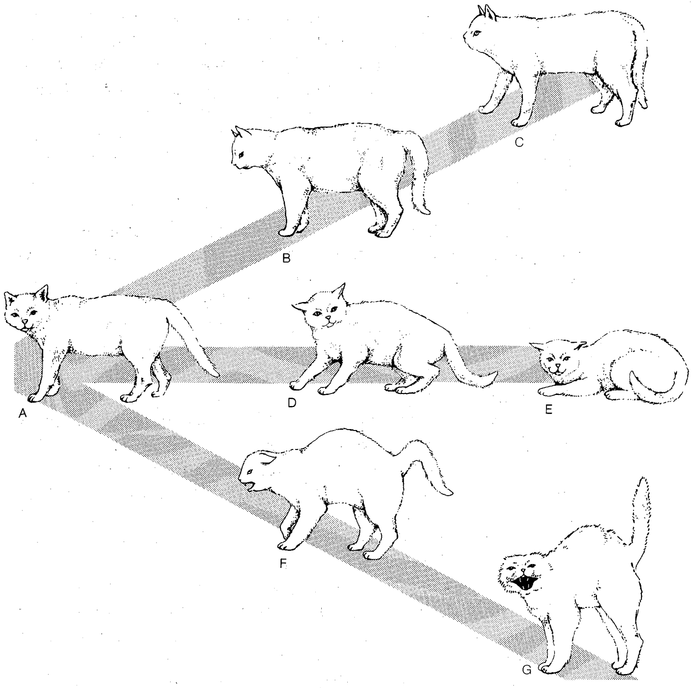

Figure 1: *Body postures during different motivational states.* (A,B,C) Increasing offensive display. (A,D,E) Increasing passive submission. Cat may roll over in E into defensive posture. (A,F,G) Increasing defensive threat display.

From P. Leyhausen, *Verhaltensstudien an Katzen* (1973); Verlag Paul Parey, West Berlin and Hamburg

Purring

While lions and other big cats roar, domestic cats and other *Felis* species purr. Purring has been described as a low, continuous, rattling hum and often is interpreted as an expression of pleasure or contentment. Purring also occurs in cats that are injured and in pain, however, so that this vocalization can be seen as the cat's "mantra"—that is, as a relaxing, self-comforting sound and a friendly mood-conveying signal.

Behavioral problems. Under conditions of domestication, the cat is subject to a variety of factors that result in behaviour indicative of emotional distress and difficulty in adapting to the home environment. Some behaviours are not abnormal but are difficult for owners to accept.

The most common behaviour problem in companion cats is that they sometimes urinate and defecate outside the litter box in the house. Organic causes include feline urologic syndrome (urinary bladder inflammation and cal-

From P. Leyhausen, *Verhaltensstudien an Katzen* (1973); Verlag Paul Parey, West Berlin and Hamburg

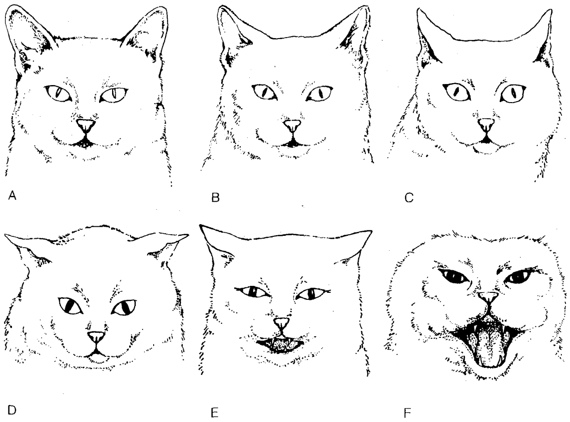

Figure 2: *Facial expressions in different motivational contexts.* Note changes in ear position and pupil dilation. (A) Alert. Offensive threat in B with fear increasing in C. Ambivalent offensive-defensive expression in D shifting to increasing threat in E and F.

culi, or stones, in the urinary tract), blocked or impacted anal glands, and constipation. Emotional causes include the addition of a new family member—another cat, a child, or a spouse. Such changes may make the cat feel insecure, so that it deposits urine and feces around the house, possibly as territorial marks for security. Cats are creatures of habit, and any change in the family structure or in daily routines—resulting, for example, from a move or even from rearranging furniture—can be stressful.

Another common behaviour problem in cats is their natural desire to rake objects such as drapes and furniture with their claws. Surgical removal of the front claws to prevent property damage is normally repugnant to cat lovers. Cats can be trained to use carpeted scratching posts in the house to satisfy this behavioral need, which may be a combination of claw cleaning and sharpening and of territorial marking.

Many cats engage in social licking and in the grooming of their feline and human companions, which is a natural display of affection and dependence. Some also engage in nursing behaviour, sucking on people's fingers and earlobes, on their own paws and tails, and on blankets and woolen clothing. Nursing may be a cat's way of regressing and relaxing into kittenish behaviour. It is often more intense in cats weaned too early or in those malnourished in kittenhood. For various emotional reasons some cats may groom themselves to the point of self-mutilation or become compulsive wool suckers and eaters.

Pica—a hunger for nonnutritive substances—may be a symptom of the need for more roughage in the diet or of feline leukemia or other health problems. As with the dog, excessive eating and drinking is frequently associated with endocrine diseases such as diabetes and thyroid dysfunction. Cats often vomit soon after eating, which is most often caused by the accumulation of fur balls in the stomach, although a food allergy, feline leukemia, or other organic cause may be involved.

Active and healthy cats often race through the house as though they were crazed. These "evening crazies" (which can also erupt early in the morning) result from the cat's ancient rhythm of actively hunting around dawn and dusk. In the domestic environment, this normal, instinctive behaviour often still occurs, to the consternation of some owners who fear that their cat may have rabies, a brain tumour, or an unstable personality.

Changes in animals' behaviour should not, therefore, be dismissed as psychological (or as simple disobedience, as when a cat suddenly becomes unhousebroken, for example), since there may be an underlying physical cause. Nevertheless, abnormal behaviour in animals often does have a nonphysical, psychological, or emotional origin, which should always be considered in the diagnosis and treatment of the ailments of companion animals.

Sleep patterns

Other traits. The cat's sleep patterns are different from those of dogs and humans. Dogs and humans have long periods of REM (rapid eye movement) sleep, the stage that is associated with dreaming. In contrast, the cat rarely lapses into REM sleep. Instead, it has a lighter, episodic sleep pattern that enables it to rest but to be instantly alert. When sick, cats have a tendency to withdraw and become inactive, which helps them conserve energy. A sick cat may seem lifeless but recover after a few days of withdrawal, which is one reason cats are said to have nine lives. (A sick cat should always be taken to a veterinarian, however; it is negligent simply to let nature take its course.)

Cats are known to have traveled hundreds of miles to find their owners in new homes to which they themselves have never been. Dogs have also performed such feats of so-called psi (psychic) trailing. Scientists have not been able to find a physiological or psychological explanation for this ability.

CATS AS PETS

The popularity of the cat, especially of pedigreed breeds, has continued to grow. The cat's independent personality, grace, cleanliness, and subtle displays of affection have wide appeal. Typically, cats are creatures of habit; they are inquisitive, but not adventurous, and are easily upset by sudden changes of routine. The ideal household cat has been separated from its mother between the ages of two and four months, raised in a clean home, kept away from unhealthy animals, and inoculated against common infectious cat diseases. Although cats often enjoy the company of other cats, especially when raised together from kittenhood, introducing a strange cat to other cats in the home can cause stress, aggression, and other behaviour problems. Cats are generally less sociable than dogs, who more readily accept a new pack member.

A good disposition and good health are important criteria for choosing a cat. Disposition varies only slightly between male and female cats. There are, however, distinct differences in disposition among the various pedigreed varieties; the Siamese, for example, is vocal and demanding, while the Persian is quiet and fastidious. The mixed breed, or "alley" cat, is a heterogeneous breed of unknown lineage; therefore, its disposition is difficult to assess. By chance, the mixed breed may prove a happier and healthier pet than a pedigreed one. On the other hand, the behaviour and vigour of the direct ancestors of pedigreed cats are indicative of the characteristics the offspring will possess as adults. But, as with the propagation of purebred dogs, the proliferation of pedigreed cats has resulted in an increase in inherited diseases, a major reason many people prefer mongrels or mixed breeds.

NUTRITION

Domestic cats should have a diet similar to that of their wild relatives. They are adapted by nature to be flesh eaters, as is shown by their alimentary tract and their dentition. The cat uses its canines to catch and kill prey, the molars to cut it up. Lack of flat-surfaced teeth prevents it from chewing or gnawing. The cat has a short intestine, and its stomach secretes digestive juices that act primarily on meat. Cats, however, like all meat-eating animals, chew grass and other plants occasionally, and small quantities of vegetables may serve as both a laxative and hairball remover.

Nutritional problems

As the science of feline nutrition has improved, specially formulated commercial diets have been developed. Serious health problems may arise when cats are fed unbalanced diets that consist exclusively of meat or tuna or when they are fed low-quality protein diets deficient in taurine. A deficiency of taurine can cause blindness and heart disease. Prescription diets are available for cats suffering from obesity, diabetes, kidney and heart disease, and urinary bladder calculi. Many cats are allergic to cow's milk, because of its high lactose content, and some are allergic to fish.

REPRODUCTION

Domestic cats reach reproductive age between 7 and 12 months. A breeding female (called a queen) can be in heat, or estrus, as many as five times a year. During these periods, which last about five days, the cat "calls," or caterwauls, intermittently. The gestation period for cats averages 63 to 65 days, and birth usually lasts about two hours. The birth is often called kittening, and the kittens are called a litter. The average litter numbers four; however, the Abyssinian usually has fewer, the Siamese more.

Each kitten is born in a separate amniotic sac that is generally broken open at the moment of birth. If it is not, the mother breaks it. She also severs the umbilical cord and eats the placenta (which in many cases stimulates lactation). The kittens are born blind, deaf, and helpless, as are many other carnivores; their senses begin to function 10 or 12 days after birth. Soon after birth the mother licks her kittens; this action cleans them and helps stimulate their circulation. Kittens at birth lack distinctive colouring, and many do not acquire their characteristic markings and colour for weeks. For example, Siamese kittens are white at birth, while blue Persians have tabby markings and black Persians are brown.

Unlike wild cats that breed once a year, the domestic cat is capable of bearing up to three litters every year. Traditionally, regulation of the cat population was accomplished by the selective killing of the newborn. In modern times, however, sterilization—by means of relatively safe and simple operations known as spaying, neutering, or altering—has become common in affluent societies. Neutering is also viewed as an adaptive measure for indoor life.

Castration of the male, ideally around six or seven months of age, helps control the adult male's tendency to "spray"—to mark objects in and around the house with his own urine. Spaying the female may help reduce the incidence of breast cancer in addition to eliminating uterine diseases and unwanted litters.

Neutered cats live longer than nonneutered ones, partly because they have less desire to roam. The average life expectancy for the domestic cat is 10 to 15 years; the oldest cat on record attained the age of 34 years.

DISEASES AND PARASITES

For many years cat treatments were simply extensions of those given dogs. Now, however, cat disorders of the skin, the eyes, the ears, the various systems (circulatory, respiratory, urinary, digestive, nervous, skeletal), and the blood, as well as contagious cat diseases and external and internal parasites, are studied, so that appropriate preventions and treatments can be developed.

Symptoms

Many cats die because their ailments become serious before their general conditions change sufficiently to reveal symptoms of illness. On the other hand, many symptoms used in diagnosing cat ailments are not definitive for given disorders. For example, signs of illness include general symptoms such as a dull coat, lack of appetite, and listlessness. Diarrhea may be a result of serious illness or simply reflect a change in diet. Tearing of the eyes, especially when accompanied by sneezing, may indicate conjunctivitis or a cold. Since, however, sneezing is the cat's only mechanism for blowing its nose, not all sneezing indicates illness. Open sores, usually at the base of the ear, around the mouth, or on the toes, can point to an ear mite or a ringworm infection or to a fight with another animal.

Cats are attacked by several kinds of external and internal parasites. External parasites are most generally found in kittens, although they can occur in adults. The most frequent parasites are fleas, but lice, ticks, and ear mites also occur. Internal parasites include roundworms, tapeworms, and protozoan coccidia. Modern veterinary medicine has made all of these easy to control.

Panleucopenia, often called feline distemper, is the best-known viral disease in cats. Highly contagious, with a high mortality rate, it is seen most often in young cats. Vaccines are effective protective measures. Rabies is less of a problem with cats than with dogs, but all free-roaming cats should be vaccinated. Vaccines have also been developed for other feline diseases, including feline leukemia, pneumonitis (chlamydiosis), viral rhinotracheitis (cat influenza), and calicivirus infections.

Cats that are permitted to wander freely outdoors are exposed to a variety of health hazards, including accidents, attacks from other animals, poisoning, fleas and other parasites, and contagious diseases such as feline acquired immune deficiency syndrome (AIDS). Cats that kill and eat rodents and other small animals can become infected with the parasitic protozoan *Toxoplasma gondii,* which can be transmitted to humans and causes the disease toxoplasmosis.

GENETICS

Cats have never been bred for economic purposes; their matings are extremely difficult to control unless the animals are completely confined. There has been relatively little scientific breeding of cats, and the facts of inheritance in these animals are not well known.

Cats are genetically far less plastic than dogs and therefore have not offered the same opportunities to breeders. The size differences between breeds in the domestic dog have no parallel in the domestic cat, nor has anything even remotely approaching the wide range of head shapes and body proportions in different breeds of dogs ever appeared among the cats. In cats the physical differences between one breed and another are largely differences in colour and texture of the coat.

Dominance of white

Colour. The most common basic coat colours in nonpedigreed cats are black, yellow, and white, with various combinations and dilutions of these colours, such as tabby or striped. Solid white is dominant to all other colours. The mode of inheritance of black is imperfectly understood, but the possibility that it is sex-linked is widely accepted. Tortoiseshell coloration, a piebald pattern of black, red, and cream that results from crossing black and yellow, is a sex-linked trait; male tortoiseshells have one extra X chromosome, resulting in an abnormal XXY chromosomal pattern. Hence, male tortoiseshells are born only rarely and are usually sterile. Similarly, tortoiseshell-and-white cats (in North America sometimes called calicoes) are almost always female.

Siamese dilution, the typical coloration of Siamese cats, has been described as a case of imperfect albinism and has been compared to the Himalayan pattern in rabbits, but its heredity is not well understood. There are also dilutions of the other ordinary colours: blue is dilute black and cream is dilute yellow. White spotting also occurs and is dominant to uniform colour.

Other characters. The long-haired coat of the Persian appears to be a simple unit character. It is recessive to short hair. Eye colour is known to be inherited, but its mode of inheritance is not thoroughly understood. Blue eye colour seems to be associated with dilution in coat colour; blue-eyed white cats are usually deaf, a fact commented on by Charles Darwin. Asymmetry of eye colour is inherited. Polydactylism, the presence of extra toes, is inherited and behaves as a dominant to the normal condition. It seems to be due to a single gene. The extra toes occur on the inner, or thumb, side of the foot.

BREEDS

The number of recognized show breeds that have defined, inherited characteristics has increased dramatically since the late 1950s as cats have become more popular as home companions. The 30 to 40 distinctive breeds can be grouped into two general categories: the long-haired Persian and the domestic shorthair. Both of these breeds occur in various subcategories based on their coat colour, such as white, cream, chinchilla (or silver), smoke, tortoiseshell, and tabby (red, blue, and so on). Other distinctive and popular breeds include the Siamese (with seal point, blue point, chocolate point, and lilac point colour variations), the long-haired Himalayan, which resembles the Siamese in coloration, and the Abyssinian, Burmese, Manx, rex, and Russian blue. (M.W.Fo./Ed.)

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Cells: Their Structures and Functions

As the smallest units retaining the fundamental properties of life, cells are the "atoms" of the living world. A single cell is often a complete organism in itself, such as a bacterium or yeast. Other cells, by differentiating in order to acquire specialized functions and cooperating with other specialized cells, become the building blocks of large multicellular organisms as complex as the human being. Although they are much larger than atoms, these building blocks are still very small. The smallest known cells are a group of tiny bacteria called mycoplasmas; some of these single-celled organisms are spheres about 0.3 micrometre in diameter, with a total mass of 10^{-14} gram—equal to that of 8,000,000,000 hydrogen atoms. Human cells typically have a mass 400,000 times larger, but even they are only about 20 micrometres across. It would require a sheet of about 10,000 human cells to cover the head of a pin, and each human being is composed of 10,000,000,000,000 cells.

This article discusses the cell both as an individual unit and as a contributing part of a larger organism. As an individual unit the cell is capable of digesting its own nutrients, providing its own energy, and replicating itself in order to produce succeeding generations. It can be viewed as an enclosed vessel composed of even smaller units that serve as its skin, skeleton, brain, and digestive tract. Within this vessel innumerable chemical reactions take place simultaneously, all of them controlled so that they contribute to the life and procreation of the cell. As part of a multicellular organism the cell keeps in constant communication with its neighbours. As it receives nutrients from and expels wastes into its surroundings, it adheres to and cooperates with other cells. Cooperative assemblies of similar cells form tissues, and a cooperation between tissues in turn forms organs, the functional units of an organism.

Special emphasis is given in this article to animal cells, with some discussion of the energy-synthesizing processes and extracellular components peculiar to plants. For detailed discussion of the biochemistry of plant cells, see PHOTOSYNTHESIS. The larger units to which cells contribute are discussed in TISSUES AND FLUIDS; and in such specialized articles as MUSCLES AND MUSCLE SYSTEMS. BIOCHEMICAL COMPONENTS OF ORGANISMS gives full information on the compounds making up the cell, and the vital reactions involving these compounds are discussed in METABOLISM. For full-length treatment of the genetic events in the cell nucleus, see GENETICS AND HEREDITY, THE PRINCIPLES OF.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 321, 322, 323, and 331. (B.M.A.)

The article is divided into the following sections:

The nature and function of cells

Specialized compartments, or organelles

A cell is enclosed by the plasma membrane, which forms a selective barrier allowing nutrients to enter and waste products to leave. The interior is organized into many specialized compartments, or organelles, each surrounded by a separate membrane. One major organelle, the nucleus, contains the genetic information necessary for cell growth and reproduction. Each cell contains only one nucleus; other types of organelles are present in multiple copies in the cellular contents, or cytoplasm. The mitochondria are responsible for the energy transactions necessary for cell survival. The lysosomes digest unwanted materials within the cell. The endoplasmic reticulum and the Golgi apparatus play an important role in the internal organization of the cell by synthesizing selected molecules and then processing, sorting, and directing them to their proper locations. (Plant cells, in addition to all the above organelles, contain chloroplasts, which are responsible for photosynthesis, whereby the energy of sunlight is used to convert molecules of carbon dioxide [CO_2] and water [H_2O] into carbohydrates.) Between all of these organelles is the space in the cytoplasm called the cytosol, which is organized around a framework of fibrous molecules constituting the cytoskeleton. The cytosol contains more than 10,000 different kinds of molecules involved in cellular biosynthesis.

These specialized compartments characterize the general class of cells known as eukaryotes. The other general class, the prokaryotes, are simpler and generally smaller than eukaryotic cells, and they lack a defined nucleus as well as the other organelles. However, all cells share a strong similarity in biochemical function.

THE CELL AS A SELF-REPLICATING COLLECTION OF CATALYSTS

From a chemical point of view, a cell is a special collection of molecules, enclosed by a membrane, that has the ability to reproduce itself from the molecules in its surroundings.

Catalytic macromolecules

This reproductive process occurs in two steps: cell growth and cell division. In the first step, the cell ingests certain simple molecules from its surroundings by selectively transferring them through its plasma membrane. Once inside the cell, these molecules are immediately subjected to the action of a large number of highly specialized catalysts, substances that speed up chemical reactions without undergoing permanent change themselves. The catalysts act by binding their surfaces to molecules in specific ways that increase the molecules' reactivity. In a living cell, each catalyst is a large, elaborately folded molecule (called a macromolecule) composed of thousands of precisely arranged atoms. The sculpted surface of the macromolecule binds tightly to one or more small molecules, distorting the chemical bonds between their atoms in a way that speeds up a particular chemical reaction—typically increasing its rate more than a millionfold. These sophisticated catalytic macromolecules are found only in the biologic world, and nothing outside of cells even remotely approaches their specificity of action.

When researchers follow specific molecules with radioactive tracers, they discover that the biologic catalysts create chains of reactions. In other words, the reaction catalyzed by one macromolecule produces the starting material, or substrate, of a second catalyst, and so on. In this way, the catalytic macromolecules use the small molecules brought into the cell from the outside environment to create increasingly complex reaction products (see Figure 1). These final products are a second set of biologic macromolecules identical to the original set in the cell. Once the number of molecules of each type has doubled, the cell goes through step two in the reproductive process, dividing to create two daughter cells, each identical to the original parent cell. Through many such cycles of cell growth and division, each parent cell can give rise to millions of daughter cells, in the process converting large amounts of inanimate matter into biologically active molecules.

The structure of biologic catalysts. Because the carbon atom can form stable bonds with four other atoms, it is uniquely suitable for the construction of complex chain-and-ring molecules. These molecules typically contain hydrogen, oxygen, and nitrogen atoms as well as carbon, and they consist of anywhere from 10 to millions of atoms linked together in specific arrays. Cells are largely composed of these carbon compounds, which form the basis of the field of organic chemistry. Most, but not all, of the carbon-containing molecules in cells are built up from members of one of four different families of small organic molecules: sugars, amino acids, nucleotides, and fatty acids. Each of these families contains a group of molecules that resemble one another in both structure and function. An example of one small molecule in each family is shown in Figure 2. In addition to other important functions, these molecules are used in the construction of larger assemblies: the sugars can be linked to form polysaccharides such as starch and glycogen, the amino acids to form proteins, the nucleotides to form the DNA (deoxyribonucleic acid) and RNA (ribonucleic acid) of the chromosomes, and the fatty acids to form the phospholipids of all cell membranes.

From B. Alberts *et al.*, *Molecular Biology of The Cell* (1983); Garland Publishing, Inc.

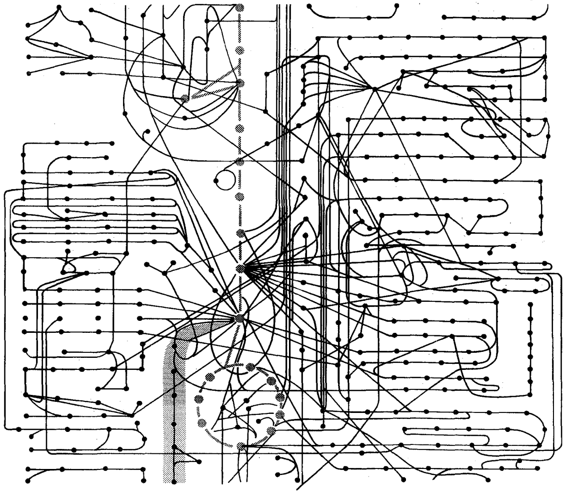

Figure 1: *Some of the chemical reactions occurring in a cell.* Radiating from the glycolytic pathway and tricarboxylic acid cycle are about 500 common metabolic reactions.

From B. Alberts *et al.*, *Molecular Biology of The Cell* (1983); Garland Publishing, Inc.

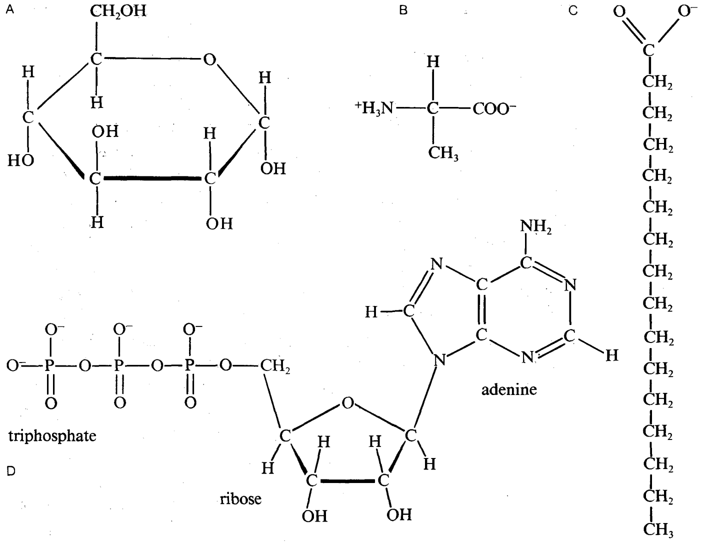

Figure 2: (A) The structure of glucose, a sugar. (B) The amino acid alanine. (C) Palmitic acid, a fatty acid. (D) The chemical structure of adenosine triphosphate (ATP), a nucleotide.

Table 1 lists the approximate proportions of each type of molecule found in a typical cell. Aside from water, which forms 70 percent of its mass, a cell is composed mostly of macromolecules (*e.g.*, proteins, RNA, DNA, and polysaccharides). By far the largest portion of macromolecules are the proteins.

Table 1: Approximate Chemical Composition of a Typical Mammalian Cell

component	percent of total cell weight
H_2O	70
Inorganic ions (Na^+, K^+, Mg^{2+}, Ca^{2+}, Cl^-, etc.)	1
Miscellaneous small metabolites	3
Proteins	18
RNA	1.1
DNA	0.25
Phospholipids and other lipids	5
Polysaccharides	2

Composition and structure of enzymes

Most of the catalytic macromolecules in cells are enzymes, which are proteins composed of 20 different amino acids joined end to end in specific sequences. An average-sized protein macromolecule contains a string of about 400 amino acid molecules. Each amino acid has a different side chain of atoms that gives it a distinct chemical property. The side chains interact with one another in specific ways, usually causing the protein molecule to fold into a compact globular form. In theory, nearly an infinite variety of proteins can be formed, each with a different sequence of amino acids. However, nearly all of these proteins would fail to fold up in the unique way required to form an efficient catalytic surface and would therefore be useless in the growth and division of a cell. The proteins actually present in a cell are products of a long evolutionary history, during which the ancestor proteins were naturally selected for their ability to fold into specific three-dimensional forms useful for cell survival. During this process of evolution, each protein developed a unique surface with a highly specific set of functions.

A second type of catalytic macromolecule in cells are RNA molecules, which consist of linear chains of four different nucleotides. RNA molecules resemble proteins in being able to fold in specific ways to form unique surfaces. As with proteins, the specific sequence of nucleotide subunits in an RNA chain gives each macromolecule a unique character. RNA molecules are much less frequently used as catalysts in cells than are protein molecules, presumably because proteins, with the greater chemical variety of their 20 different amino acid side chains, are more powerful catalysts. However, RNA molecules are thought to have preceded protein molecules during evolution and to have catalyzed most of the chemical reactions required before cells could evolve (see below *The evolution of cells*).

Energy in chemical reactions

Coupled chemical reactions. Cells must obey the laws of chemistry and thermodynamics. When two molecules react with each other inside a cell, their atoms are rearranged, forming different molecules as reaction products and releasing energy in the process. Such chemical reactions can occur only in one direction; that is, the reaction product molecules cannot spontaneously react, in a reversal of the original process, to reform the original molecules. This directionality of chemical reactions is explained by the fact that molecules only change from states of higher free energy to states of lower free energy. Free energy is the ability to perform work (in this case, the "work" is the rearrangement of atoms in the chemical reaction). When work is performed, some free energy is used and lost, with the result that the process ends at lower free energy. To use a familiar mechanical analogy, water at the top of a hill has the ability to perform the "work" of flowing downhill (*i.e.,* it has high free energy), but once it has flowed downhill it cannot flow back up (*i.e.,* it is in a state of low free energy). However, through another work process—that of a pump, for example—the water can be returned to the top of the hill, thereby recovering its ability to flow downhill. In thermodynamic terms, the free energy of the water has been increased by energy from an outside source. In the same way, the product molecules of a chemical reaction in a cell cannot reverse the reaction and return to their original state unless energy is supplied by coupling the process to another chemical reaction.

All catalysts, including enzymes, accelerate chemical reactions without affecting their direction. To return to the mechanical analogy, enzymes cannot make water flow uphill, although they can provide specific pathways for a downhill flow. Yet most of the chemical reactions that the cell needs to synthesize new molecules necessary for its growth require an uphill flow. In other words, the reactions require more energy than their starting molecules can provide.

Cells use a single strategy over and over again in order to get around the limitations of chemistry: they use the energy from an energy-releasing chemical reaction to drive an energy-absorbing reaction that would otherwise not occur. A useful mechanical analogy might be a mill wheel driven by the water in a stream. The water, in order to flow downhill, is forced to flow past the blades of the wheel, causing the wheel to turn. In this way, part of the energy from the moving stream is harnessed to move a mill wheel, which may be linked to a winch. As the winch turns, it can be used to pull a heavy load uphill. Thus, the energy-absorbing (but useful) uphill movement of a load can be driven by coupling it directly to the energy-releasing flow of water.

In cells, enzymes play the role of mill wheels by coupling energy-releasing reactions with energy-absorbing reactions. As discussed below, in cells the most important energy-releasing reaction serving a role similar to that of the flowing stream is the hydrolysis of adenosine triphosphate (ATP). In turn, the production of ATP molecules in the cells is an energy-absorbing reaction that is driven by being coupled to the energy-releasing breakdown of sugar molecules. In retracing this chain of reactions, it is necessary first to understand the source of the sugar molecules.

Photosynthesis: the beginning of the food chain. Sugar molecules are some of the carbohydrates produced by the process of photosynthesis in plants and certain bacteria. These organisms lie at the base of the food chain, in that animals and other nonphotosynthesizing organisms depend on them for a constant supply of life-supporting organic molecules. Humans, for example, obtain these molecules by eating plants or other organisms that have previously eaten food derived from photosynthesizing organisms.

Conversion of solar energy to chemical bond energy

Plants and photosynthetic bacteria are unique in their ability to convert the freely available electromagnetic energy in sunlight into chemical bond energy, the energy that holds atoms together in molecules and is transferred or released in chemical reactions. The process of photosynthesis can be summarized by the following equation:

$$\text{(solar) energy} + CO_2 + H_2O \quad \rightarrow \quad \text{sugar molecules} + O_2.$$

The energy-absorbing photosynthetic reaction is the reverse of the energy-releasing oxidative decomposition of sugar molecules. During photosynthesis, chlorophyll molecules absorb energy from sunlight and use it to fuel the production of simple sugars and other carbohydrates. The resulting abundance of sugar molecules and related biological products makes possible the existence of nonphotosynthesizing life on Earth.

ATP: fueling chemical reactions. Enzymes break down organic foodstuffs into sugars and other small molecules such as amino acids and fatty acids. Some of these small molecules serve as building blocks for the production of the cell's macromolecules, but more than half of them are further degraded in the cytoplasm in order to provide the cell with ATP. ATP, the common carrier of energy inside the cell, is made from adenosine diphosphate (ADP) and inorganic phosphate (P_i; the structure of the molecule is shown in Figure 2). Stored in the chemical bond holding the terminal phosphate compound onto the ATP molecule is the energy derived from the breakdown of sugars. The removal of the terminal phosphate, through the water-mediated reaction called hydrolysis, releases this energy, which in turn fuels a large number of crucial energy-absorbing reactions in the cell. Hydrolysis can be summarized as follows:

$$ATP + H_2O \quad \rightarrow \quad ADP + P_i + \text{energy}.$$

The formation of ATP is the reverse of this equation, requiring the addition of energy. It begins with glycolysis, a form of fermentation in which glucose is broken down step by step in a series of nine enzymatic reactions, each successive reaction involving an intermediate sugar containing phosphate. In the process, the six-carbon glucose is converted into two molecules of the three-carbon pyruvic acid. Some of the energy released through the glycolysis of each glucose molecule is absorbed in the formation of two ATP molecules.

The transfer of chemical bond energy

The second stage in the breakdown of sugars is a set of interrelated reactions called the tricarboxylic acid cycle. This cycle takes the three-carbon pyruvic acid produced in glycolysis and uses its carbon atoms to form CO_2 while transferring its hydrogen atoms to special carrier molecules, where they are held in high-energy linkage.

In the third and last stage in the breakdown of sugars, oxidative phosphorylation, the high-energy hydrogen atoms are first separated into protons and high-energy electrons. The electrons are then passed from one electron carrier to another by means of an electron-transport chain. Each electron carrier in the chain has an increasing affinity for electrons, with the final electron acceptor being molecular oxygen (O_2). As separated electrons and protons, the hydrogen atoms are transferred to O_2 to form water. This reaction releases a large amount of energy, which drives the synthesis of a large number of ATP molecules from ADP and P_i. (For further discussion of the electron-transport chain, see below *The mitochondrion and the chloroplast: Metabolic functions.*)

Most of the cell's ATP is produced when the products of glycolysis are oxidized completely by a combination of the tricarboxylic acid cycle and oxidative phosphorylation, glycolysis alone producing relatively small amounts. Glycolysis is an anaerobic reaction; that is, it can occur even in the absence of oxygen. The tricarboxylic acid cycle and oxidative phosphorylation, on the other hand, require oxygen. Glycolysis forms the basis of anaerobic fermentation, and it presumably was a major source of ATP for early life on Earth, when very little oxygen was available in the atmosphere. Eventually, however, bacteria evolved that were able to carry out photosynthesis. Photosynthesis liberated these bacteria from a dependence on the metabolism of organic materials that had accumulated from natural processes, and it also released oxygen into the atmosphere. Over a prolonged period of time, the concentration of molecular oxygen increased until it became freely available in the atmosphere. The aerobic tricarboxylic acid cycle and oxidative phosphorylation then evolved, enabling sugars to be broken down completely to carbon dioxide and water. The resulting aerobic cells made much more efficient use of foodstuffs than their anaerobic ancestors, because they could convert much larger amounts of chemical bond energy into ATP.

THE CELL AS A REPLICATOR OF INFORMATION

Cells can thus be seen as a self-replicating network of catalytic macromolecules engaged in a carefully balanced series of energy conversions that drive biosynthesis and cell movement. But energy alone is not enough to make self-reproduction possible; the cell must contain detailed instructions that dictate exactly how that energy is to be used. These instructions are analogous to the blueprints that a builder uses to construct a house; in the case of cells, however, the blueprints themselves must be duplicated along with the cell before it divides, so that each daughter cell can retain the information that it needs for its own replication. This information constitutes the cell's heredity.

DNA: the genetic material. During the early 19th century, it became widely accepted that all living organisms are composed of cells arising only from the growth and division of other cells. The improvement of the microscope then led to an era during which many biologists made intensive observations of the microscopic structure of cells. By 1885 a substantial amount of indirect evidence indicated that chromosomes—dark-staining threads in the cell nucleus—carried the information for cell heredity. It was later shown that chromosomes are about half DNA and half protein by weight.

Self-duplication of DNA

In the simplest possible case, each macromolecule in a cell would contain the information for its own duplication. Self-duplication is theoretically feasible, for example, for the nucleic acids in a cell. DNA in particular has a two-stranded polynucleotide structure that could separate into single-stranded chains. Each strand could then form another two-stranded complex by guiding the synthesis of a second polynucleotide chain with a complementary nucleotide sequence. The final products would be two double-stranded molecules, each identical to the parent molecule. The revolutionary discovery suggesting that DNA molecules could provide the information for their own replication came in 1953, when researchers James Watson and Francis Crick proposed a model for the structure of the double-stranded DNA molecule (called the DNA double helix). This scheme, in which each strand serves as a template in the synthesis of a complementary strand, is outlined in Figure 3. Subsequent research showed that DNA actually replicates in this way and that it carries the genetic information for reproduction of the entire cell.

From J.D. Watson *et al., The Molecular Biology of the Gene*, vol. 1, p. 283, 4th ed. (1987); The Benjamin/Cummings Publishing Co.

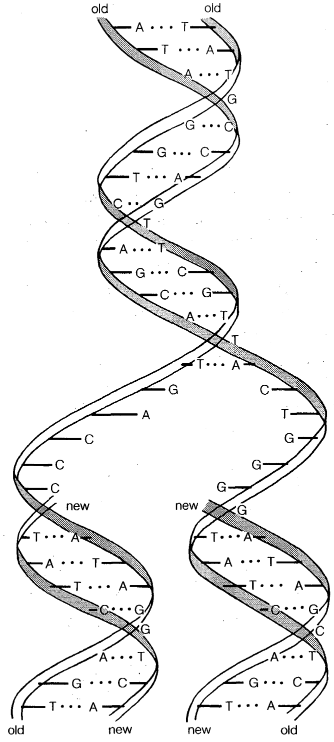

Figure 3: The initial proposal of the structure of DNA by James Watson and Francis Crick, which was accompanied by a suggestion on the means of replication.

All of the genetic information in a cell was initially thought to be confined to the DNA in the chromosomes of the cell nucleus. It is now known that small amounts of additional genetic information are present in the DNA of much smaller chromosomes located in two types of organelles in the cytoplasm. These organelles are the mitochondria in animal cells and the mitochondria and chloroplasts in plant cells. The special chromosomes carry the information coding for a few of the many proteins and RNA molecules needed by the organelles. They also hint at the evolutionary origin of these organelles, which are thought to have originated as free-living bacteria that were taken up by symbiosis.

RNA: replicated from DNA. It is possible for RNA to replicate itself by mechanisms related to those used by DNA, even though it has a single-stranded instead of a double-stranded structure. In early cells RNA is thought to have replicated itself in this way. However, all of the RNA in present-day cells is synthesized by special enzymes that construct a single-stranded RNA chain by using one strand of the DNA helix as a template. Although RNA molecules are synthesized in the cell nucleus, where the DNA is located, most of them are transported to the cytoplasm before they carry out their functions.

Replication of proteins by messenger RNA

The RNA molecules in cells have two main roles. Some fold up in ways that allow them to serve as catalysts for specific chemical reactions. Others serve as "messenger RNA," which provides templates specifying the synthesis of proteins. Ribosomes, tiny protein-synthesizing machines located in the cytoplasm, "read" the messenger RNA molecules and "translate" them into proteins by using the genetic "code." In this translation, the sequence of nucleotides in the messenger RNA chain is decoded three nucleotides at a time, and each nucleotide triplet (called a codon) specifies a particular amino acid. Thus, a nucleotide sequence in the DNA specifies a protein provided that a messenger RNA molecule is produced from that DNA sequence. Each region of the DNA sequence specifying a protein in this way is called a gene.

By the above mechanisms, DNA molecules catalyze not only their own duplication but also that of all protein molecules. A single human cell contains about 10,000 different proteins produced by the expression of 10,000 different genes. Actually, a set of human chromosomes is thought to contain DNA with enough information to express between 30,000 and 100,000 proteins, but most of these proteins seem to be made only in specialized types of cells and are therefore not present throughout the body. (For further discussion, see below *The nucleus.*)

THE CELL AS AN ORGANIZED UNIT

Intracellular communication. A cell with its many different DNA, RNA, and protein molecules is quite different from a test tube containing the same components. When a cell is dissolved in a test tube, thousands of different types of molecules randomly mix together. In the living cell, however, these components are kept in different places, reflecting the high degree of organization essential for the growth and division of the cell. Maintaining this internal organization requires a continual input of energy, because spontaneous chemical reactions always create disorganization. Thus, some of the energy released by ATP hydrolysis fuels processes that organize macromolecules inside the cell.

When a eukaryotic cell is examined at high magnification in an electron microscope, it becomes apparent that specific membrane-bounded organelles divide the interior into a variety of subcompartments. (The proportions of the total cell volume occupied by these compartments are shown in Table 2.) Although not detectable in the electron microscope, it is clear from biochemical assays that each organelle contains a different set of macromolecules. This biochemical segregation reflects the functional specialization of each compartment. Thus, the mitochondria, which produce most of the cell's ATP, contain all of the enzymes needed to carry out the tricarboxylic acid cycle and oxidative phosphorylation. Similarly, the degradative enzymes needed for the intracellular digestion of unwanted macromolecules are confined to the lysosomes.

Table 2: The Relative Volumes Occupied by Some Cellular Compartments in a Typical Liver Cell

cellular compartment	percent of total cell volume	approximate number per cell
Cytosol	54	1
Mitochondrion	22	1,700
Endoplasmic reticulum plus Golgi apparatus	15	1
Nucleus	6	1
Lysosome	1	300

The molecular sorting process

It is clear from this functional segregation that the many different proteins specified by the genes in the cell nucleus must be transported to the compartment where they will be used. Not surprisingly, the cell contains an extensive membrane-bound system devoted to maintaining just this intracellular order. The system serves as a post office, guaranteeing the proper routing of newly synthesized macromolecules to their proper destinations. The molecular details of the many necessary sorting operations are not known, but the sorting of proteins, the most abundant molecules in cells, has been extensively studied.

All proteins are synthesized on ribosomes located in the cytosol. As soon as the first portion of the amino acid sequence of a protein emerges from the ribosome, it is inspected for the presence of a short "endoplasmic reticulum (ER) signal sequence." Those ribosomes making proteins with such a sequence are transported to the surface of the ER membrane, where they complete their synthesis; the proteins made on these ribosomes are immediately transferred through the ER membrane to the inside of the ER compartment. Proteins lacking the ER signal sequence remain in the cytosol and are released from the ribosomes when their synthesis is completed. This chemical decision process places some newly completed protein chains in the cytosol and others within an extensive membrane-bounded compartment in the cytoplasm, representing the first step in intracellular protein sorting.

The newly made proteins in both cell compartments are then sorted further according to additional signal sequences that they contain. Some of the proteins in the cytosol remain there, while others go to the surface of mitochondria or (in plant cells) chloroplasts, where they are transferred through the internal membranes into the organelles. Subsignals on each of these proteins then designate exactly where in the organelle the protein belongs. The proteins initially sorted into the ER have an even wider range of destinations. Some of them remain in the ER, where they function as part of the organelle. Most enter transport vesicles and pass to the Golgi apparatus, separate membrane-bounded organelles that contain at least three subcompartments. Some of the proteins are retained in the subcompartments of the Golgi, where they are utilized for functions peculiar to that organelle. Most eventually enter vesicles that leave the Golgi for other cellular destinations such as the plasma membrane, lysosomes, or special secretory vesicles. (For further discussion, see below *Internal membranes.*)

Intercellular communication. Formation of a multicellular organism starts with a small collection of similar cells in an embryo and proceeds by continuous cell division and specialization to produce an entire community of cooperating cells, each with its own role in the life of the organism. Through cell cooperation, the organism becomes much more than the sum of its component parts.

Differentiation of cells with common genetic inheritance

A fertilized egg multiplies and produces a whole family of daughter cells, each of which adopts a structure and function according to its position in the entire assembly. All of the daughter cells contain the same chromosomes and therefore the same genetic information. Despite this common inheritance, different types of cells behave differently and have different structures. In order for this to be the case, they must express different sets of genes, so that they produce different proteins despite their identical embryological ancestors.

During the development of an embryo, it is not sufficient for all of the cell types found in the fully developed individual simply to be created. Each cell type must form in the right place at the right time and in the correct proportion; otherwise, there would be a jumble of randomly assorted cells in no way resembling an organism. The orderly development of an organism depends on a process called cell determination, in which initially identical cells somehow "read" their positions in relation to their neighbours and differentiate in ways appropriate for cells in their positions. This requires an elaborate system of cell-to-cell communication in early embryos.

The specification of cell differences seems to involve a series of steps. First, cells differentiate according to their positions at an early stage of development. Second, each cell "remembers" its differences and passes the genetic information to its daughter cells. A later generation of cells takes a second reading, creating additional position-dependent differences. These differences are also remembered as the cells grow and divide. By repeating this process and exploiting the ability of cells to remember their past, it is possible to construct a complex organism without requiring that cells be able to communicate with one another over long distances. In fact, the distances over which cells communicate in embryos are estimated to be only one millimetre or less (equal to the width of about 100 cells).

The molecular basis for such biologic pattern formation is not known, although many different theoretical models have been proposed. It is believed that related mechanisms persist in the adult, maintaining the correct proportions of each cell type in a tissue or organ and restoring proper form after disease or injury. (For further discussion, see below *Cell differentiation.*) (B.M.A.)

The plasma membrane

A thin membrane, some .005 micrometre across, surrounds every living cell, delimiting the cell from the environment around it. Enclosed by this plasma membrane are the cell's constituents, often large, water-soluble, highly charged molecules such as proteins, nucleic acids, carbohydrates, and substances involved in cellular metabolism. Outside the cell, in the surrounding water-based environment, are ions, acids, and alkalis that are toxic to the cell, as well as nutrients that the cell must absorb in order to live and grow. The plasma membrane, therefore, has two functions: first, to be a barrier keeping the constituents of the cell in and unwanted substances out; and second, to be a gate allowing transport into the cell of essential nutrients and movement from the cell of waste products.

The dual function of the plasma membrane

CHEMICAL COMPOSITION AND STRUCTURE OF THE MEMBRANE

Most current knowledge about the biochemical constituents of plasma membranes originates in studies of red blood cells. The chief advantage of these cells for experimental purposes is that they may be obtained easily in large amounts and that they have no typical membrane structure, other than the plasma membrane itself, to interfere with study of that structure. Careful studies of these and other cell types have shown that all plasma membranes are composed of proteins and fatty-acid-based lipids. Cell membranes actively involved in metabolism contain a higher proportion of protein; thus, the membrane of the mitochondrion, the most rapidly metabolizing organelle of the cell, contains as much as 75 percent protein, while the membrane of the Schwann cell, which forms an insulating sheath around many nerve cells, has as little as 20 percent protein.

Membrane lipids. Membrane lipids are principally of two types, phospholipids and sterols (generally cholesterol). Both types share the defining characteristic of lipids—they dissolve readily in organic solvents—but in addition they both have a region that is attracted to and soluble in water. This "amphiphilic" property (having a dual attraction, *i.e.,* containing both a lipid-soluble and a water-soluble region) is basic to the role of lipids as building blocks of cellular membranes. Phospholipid molecules have a head (often of glycerol) to which are attached two long fatty acid chains that look much like tails. These tails are repelled by water and dissolve readily in organic solvents, giving the molecule its lipid character. To another part of the head is attached a phosphoryl group with a negative electrical charge; to this group in turn is attached another group with a positive or neutral charge. This portion of the phospholipid dissolves in water, thereby completing the molecule's amphiphilic character. In contrast, sterols have a complex hydrocarbon ring structure as the lipid-soluble region, and a hydroxyl grouping as the water-soluble region.

When dry phospholipids, or a mixture of such phospholipids and cholesterol, are immersed in water under laboratory conditions, they spontaneously form globular structures called liposomes. Investigation of the liposomes shows them to be made of concentric spheres, one sphere inside of another and each forming half of a bilayered wall. A bilayer is composed of two sheets of phospholipid molecules with all of the molecules of each sheet aligned in the same direction. In a water medium, the phospholipids of the two sheets align so that their water-repellent, lipid-soluble tails are turned and loosely bonded to the tails of the molecules on the other sheet. The water-soluble heads turn outward into the water, to which they are chemically attracted. In this way, the two sheets form a fluid, sandwichlike structure, with the fatty acid chains in the middle mingling in an organic medium while sealing out the water medium.

This type of lipid bilayer, formed by the self-assembly of lipid molecules, is the basic structure of the plasma membrane of the cell. It is the most stable thermodynamic structure that a phospholipid–water mixture can take up: the fatty acid portion of each molecule dissolved in the organic phase formed by the identical regions of the other molecules and the water-attractive regions surrounded by water and facing away from the fatty acid regions. The chemical affinity of each region of the amphiphilic molecule is thus satisfied in the bilayer structure.

Membrane proteins. Membrane proteins are also of two general types. One type, called the extrinsic proteins, is loosely attached by ionic bonds or calcium bridges to the electrically charged phosphoryl surface of the bilayer. They are also attached to the second type of protein, called the intrinsic proteins. The intrinsic proteins, as their name implies, are firmly bound within the phospholipid bilayer. Many of them contain 20- to 24-membrane-long amino acid sequences—a length just sufficient to cross the bilayer—that may extend through and bind to the fatty acid internal regions of the membrane.

Membrane-bound intrinsic proteins

Many intrinsic proteins bear on their outer surfaces side chains of complex sugars, which extend into the aqueous environment around the cell. These are involved in cell-to-cell recognition (see below *The cell matrix and cell-to-cell communication*).

Membrane fluidity. One of the triumphs of molecular biology during the decade from 1965 to 1975 was the recognition of the cell membrane as a fluid collection of amphiphilic molecules. This array of proteins, sterols, and phospholipids is organized into a liquid crystal, a structure that lends itself readily to rapid cell growth. Measurements of the membrane's viscosity show it as a fluid a hundred times as viscous as water, similar to a thin oil. The phospholipid molecules diffuse readily in the plane of the bilayer, sliding over the face of the membrane as their loose chemical bonds permit. Many of the membrane's proteins also have this freedom of movement, but others are fixed in the membrane by interaction with the cell's cytoskeleton. Proteins never, and phospholipids only rarely, move from one half of the bilayer to the other. Newly synthesized phospholipids insert themselves easily into the existing plasma membrane. Intrinsic proteins are inserted during their synthesis on ribosomes bound to the inner surface of the membrane, while extrinsic proteins found on the internal surface of the plasma membrane are synthesized on free, or unattached, ribosomes, liberated into the cytoplasm, and then brought to the membrane by diffusion.

TRANSPORT ACROSS THE MEMBRANE

The chemical structure of the plasma membrane makes it remarkably flexible, the ideal boundary of rapidly growing and dividing cells. Yet the membrane is also a formidable barrier, allowing some dissolved substances, or solutes, to pass while blocking others. Lipid-soluble molecules and some small molecules can permeate the membrane, but the lipid bilayer effectively repels the many large, water-soluble molecules and electrically charged ions that the cell must import or export in order to live. Transport of these vital substances is carried out by the membrane proteins, particularly the intrinsic proteins. These molecules form a variety of transport systems: some are open channels, which allow ions to diffuse directly into the cell; others are "facilitators," which, through a little-understood chemical transformation, help solutes diffuse past the lipid screen; yet others are "pumps," which force solutes through the membrane when they are not concentrated enough to diffuse spontaneously. Particles too large to be diffused or pumped are often swallowed or disgorged whole by an opening and closing of the membrane.

Principle of diffusion

Behind this movement of solutes across the plasma membrane is the principle of diffusion. According to this principle, a dissolved substance diffuses down a concentration gradient; that is, given no energy from an outside source, it moves from a place where its concentration is

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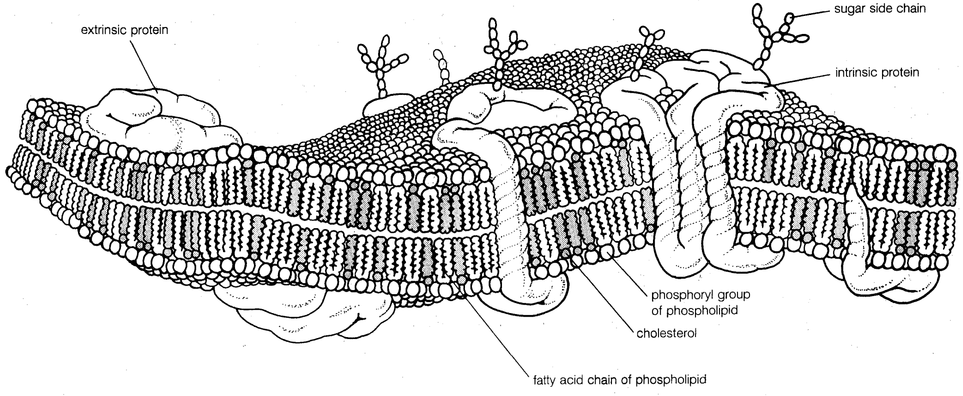

Figure 4: *A molecular view of the plasma membrane.*
Intrinsic proteins penetrate and bind tightly to the lipid bilayer; extrinsic proteins are loosely bound to the phosphoryl surface. Some intrinsic proteins present sugar side chains on the cell surface.

high to a place where its concentration is low. Diffusion continues down this gradually decreasing gradient until a state of equilibrium is reached, at which point there is an equal concentration in both places and an equal, random diffusion in both directions.

A solute at high concentration is at high free energy; that is, it is capable of doing more "work" (the work being that of diffusion) than a solute at low concentration. In performing the work of diffusion, the solute loses free energy, so that when it reaches equilibrium at a lower concentration, it is unable to return spontaneously (under its own energy) to its former high concentration. However, by the addition of energy from an outside source (through the work of an ion pump, for example), the solute may be returned to its former concentration and state of high free energy. This "coupling" of work processes is, in effect, a transferal of free energy from the pump to the solute, which is then able to repeat the work of diffusion. (See also above *The nature and function of cells: Coupled chemical reactions.*)

For most substances of biologic interest, the concentrations inside and outside the cell are different, creating concentration gradients down which the solutes spontaneously diffuse, provided they can permeate the lipid bilayer. Membrane channels and diffusion facilitators bring them through the membrane by passive transport; that is, the changes that the proteins undergo in order to facilitate diffusion are powered by the diffusing solutes themselves. For the healthy functioning of the cell, some solutes must remain at different concentrations on each side of the membrane; if through diffusion they approach equilibrium, they must be pumped back up their gradients by the process of active transport. Those membrane proteins serving as pumps accomplish this by coupling the energy required for transport to the energy produced by cell metabolism or by the diffusion of other solutes.

Permeation. Permeation is the diffusion, through a barrier, of a substance in solution. The rates at which biologically important molecules cross the plasma membrane through permeation vary over an enormous range. Proteins and sugar polymers do not permeate at all; in contrast, water and alcohols permeate most membranes in less than a second. This variation, caused by the lipid bilayer, gives the membrane its characteristic permeability. Permeability is measured as the rate at which a particular substance in solution crosses the membrane.

The barrier function of the lipid bilayer

For all cell membranes that have been studied in the laboratory, permeability increases in parallel with the permeant's ability to dissolve in organic solvents. The consistency of this parallel has led researchers to conclude that permeability is a function of the fatty acid interior of the lipid bilayer, rather than its phosphoryl exterior. This property of dissolving in organic solvents rather than water is given a unit of measure called the partition coefficient. The greater the solubility of a substance, the higher its partition coefficient, and the higher the partition coefficient, the higher the permeability of the membrane to that particular substance. For example, the water solubility of hydroxyl, carboxyl, and amino groups reduces their solubility in organic solvents and, hence, their partition coefficients. Cell membranes have been observed to have low permeability toward these groups. In contrast, lipid-soluble methyl residues and hydrocarbon rings, which have high partition coefficients, penetrate cell membranes more easily—a property useful in designing chemotherapeutic and pharmacological drugs.

For two molecules of the same partition coefficient, the one of greater molecular weight, or size, will in general cross the membrane more slowly. In fact, even molecules with very low partition coefficients can penetrate the membrane if they are small enough. Water, for example, is insoluble in organic solvents, yet it permeates cell membranes owing to the small size of its molecules. The size selectivity of the lipid bilayer is a result of its being not a simple fluid, the molecules of which move around and past a diffusing molecule, but an organized matrix, a kind of fixed grate, composed of the fatty acid chains of the phospholipids through which the diffusing molecule must fit.

Many substances do not actually cross the cell membrane through permeation of the lipid bilayer. Some electrically charged ions, for example, are repelled by organic solvents and therefore cross cell membranes with great difficulty, if at all. In these cases special holes in the membrane, called channels, allow specific ions and small molecules to diffuse directly past the bilayer.

Membrane channels. Biophysicists measuring the electric current passing through plasma membranes have found that, in general, plasma membranes have a vastly greater electrical conductance than does a membrane bilayer composed only of phospholipids and sterols. This greater conductance is thought to be conferred by the plasma membrane's proteins. A current flowing across a membrane often appears on a recording instrument as a series of bursts of various heights. These bursts represent current flowing through open channels, which are merely holes formed by intrinsic proteins traversing the lipid bilayer. No significant current flows through the membrane when no channel is open; multiple bursts are recorded when more than one channel is open.

Chemical composition of membrane channels

A rich variety of channels has been isolated and analyzed from a wide range of cell membranes. Invariably intrinsic proteins, they appear to contain numerous amino acid sequences that traverse the membrane, clearly forming a specific hole, or pore. Certain channels open and close spontaneously. Some are gated, or opened, by the chemical action of a signaling substance such as calcium, acetylcholine, or glycine, while others are gated by changes in the electrical potential across the membrane. Channels may possess a narrow specificity, allowing passage of only potassium or sodium, or a broad specificity, allowing passage of all positively charged ions (cations) or of all negatively charged ions (anions). There are channels called gap junctions that allow the passage of molecules between pairs of cells (see below *The cell matrix and cell-to-cell communication*).

The gating of channels with a capacity for ion transport is the basis of the many nerve–nerve, nerve–muscle, and nerve–gland interactions underlying neurobiological behaviour. These actions depend on the electric potential of the cell membrane, which varies with the prevailing constituents in the cell's environment. For example, if a channel that admits only potassium ions is present in a membrane separating two different potassium chloride solutions, the positively charged potassium ions tend to flow down their concentration gradient through the channel. The negatively charged chloride ions remain behind. This separation of electric charges sets up an electric potential across the membrane called the diffusion potential. The size of this potential depends on, among other factors, the difference in concentrations of the permeating ion across the membrane. A cell's plasma membrane in general contains the channels of widely different ion specificities, each channel contributing to the overall membrane potential according to the permeability and concentration ratio of the ion passing through it. Since the channels are often gated, the membrane's potential is determined by which channels are open; this in turn depends on the concentrations of signaling molecules and may change with time according to the membrane potential itself.

Most cells have about a tenfold higher concentration of sodium ions outside than inside and a reverse concentration ratio of potassium ions. Free calcium ions can be 10,000 times more concentrated outside the cell than inside. Thus, sodium-, potassium-, and calcium-selective membrane channels, by allowing the diffusion of those ions past the cell membrane and causing fluctuations in the membrane's electric potential, frequently serve as transmitters of signals from nerve cells. This diffusion through open channels, however, constantly threatens to alter the concentration of ions necessary for the cell to function. The proper distribution of ions is maintained by the action of ion pumps (see below *Primary active transport*).

Facilitated diffusion. Many water-soluble molecules that cannot penetrate the lipid bilayer are too large to fit through open channels. In this category are sugars and amino acids. Some ions, too, do not diffuse through channels. These vital substances enter and leave the cell through the

Functioning of membrane transporters

action of membrane transporters, which, like channels, are intrinsic proteins that traverse the cell membrane. Unlike channels, transporter molecules do not simply open holes in the membrane. Rather, they present sites on one side of the membrane to which molecules bind through chemical attraction. The binding site is highly specific, often fitting the atomic structure of only one type of molecule. When the molecule has attached to the binding site, then, in a process not fully understood, the transporter brings it through the membrane and releases it on the other side.

This action is considered a type of diffusion because the transported molecules move down their concentration gradients, from high concentration to low. To activate the action of the transporter, no other energy is needed than that of the chemical binding of the transported molecules. This action upon the transporter is similar to catalysis, except that the molecules (in this context called substrates) catalyze not a chemical reaction but their own translocation across the cell membrane. Two such substrates are glucose and the bicarbonate ion.

The glucose transporter. This sugar-specific transport system enables half of the glucose present inside the cell to leave within four seconds at normal body temperature. The glucose transporter is clearly not a simple membrane channel. First, unlike a channel it does not select its permeants by size, as one type of glucose is observed to move through the system a thousand times faster than its identically sized optical isomer. Second, it operates much more slowly than do most channels, moving only 1,000 molecules per second while a channel moves 1,000,000 ions. The most important difference between a membrane channel and the glucose transporter is the conformational change that the transporter undergoes while moving glucose across the membrane. Alternating between two conformations, it moves its glucose-binding site from one side of the membrane to the other. By "flipping" between its two conformational states, the transporter facilitates the diffusion of glucose; that is, it enables glucose to avoid the barrier of the plasma membrane while moving spontaneously down its concentration gradient. When the concentration reaches equilibrium, net movement of glucose ceases.

A facilitated diffusion system for glucose is present in many cell types. Similar systems transporting a wide range of other substrates (*e.g.,* different sugars, amino acids, nucleosides, ions) are also present.

The anion transporter. The best studied of the facilitated diffusion systems is that which catalyzes the exchange of anions across the red blood cell membrane. The exchange of hydroxyl for bicarbonate ions, each ion simultaneously being moved down its concentration gradient in opposite directions by the same transport molecule, is of great importance in enhancing the blood's capacity to carry carbon dioxide from tissues to the lungs. The exchange molecule for these anions is the major intrinsic protein of red blood cells; 1,000,000 of them are present on each cell, the polypeptide chain of each molecule traversing the membrane at least six times.

Secondary active transport. In some cases the problem of forcing a substrate up its concentration gradient is solved by coupling that upward movement to the downward flow of another substrate. In this way the energy-expending diffusion of the driving substrate powers the energy-absorbing movement of the driven substrate from low concentration to high. Because this type of active transport is not powered directly by the energy released in cell metabolism (see below *Primary active transport*), it is called secondary.

There are two kinds of secondary active transport: counter-transport, in which the two substrates cross the membrane in opposite directions; and co-transport, in which they cross in the same direction.

Counter-transport. An example of this system (also called antiport) begins with the sugar transporter described above. There are equal concentrations of glucose on both sides of the cell. A high concentration of galactose is then added outside the cell. Galactose competes with glucose for binding sites on the transport protein, so that mostly galactose—and a little glucose—enter the cell. The transporter itself, undergoing a conformational change, presents its binding sites for sugar at the inner face of the membrane. Here, at least transiently, glucose is in excess of galactose; it binds to the transporter and leaves the cell as the transporter switches back to its original conformation. Thus, glucose is pumped out of the cell against its gradient in exchange for the galactose riding into the cell down its own gradient.

Many counter-transport systems operate across the cell membranes of the body. A well-studied system (present in red blood, nerve, and muscle cells) pumps one calcium ion out of the cell in exchange for two or three sodium ions. This system helps maintain the low calcium concentration required for effective cellular activity. A different system, present in kidney cells, counter-transports hydrogen ions and sodium ions in a one-for-one ratio. This is important in stabilizing acidity by transporting hydrogen ions out of the body as needed.

Co-transport. In co-transport (sometimes called symport) two species of substrate, generally an ion and another molecule or ion, must bind simultaneously to the transporter before its conformational change can take place. As the driving substrate is transported down its concentration gradient it drags with it the driven substrate, which is forced to move up its concentration gradient. The trans-

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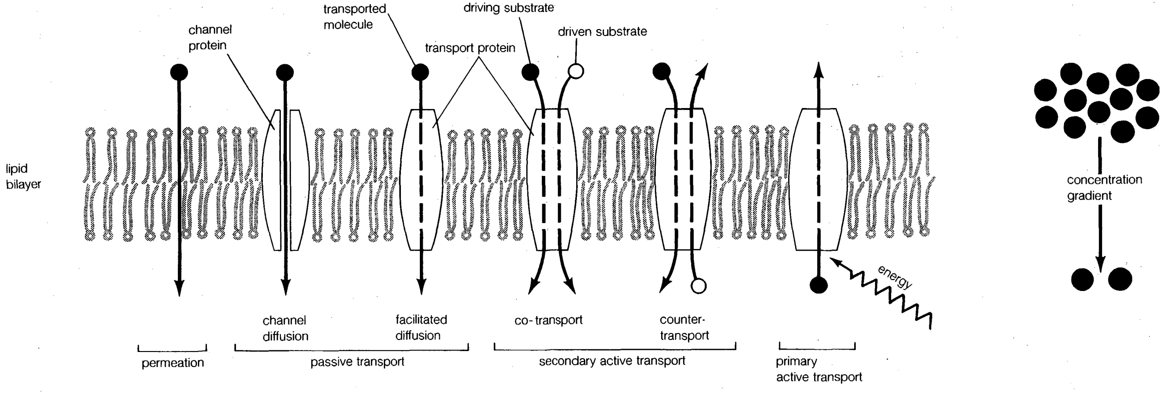

Figure 5: *The permeation and transport of molecules through the lipid bilayer of a cell membrane.*
Some small molecules can directly permeate the lipid bilayer. Many charged molecules, repelled by the bilayer, diffuse through membrane channels. Other charged or large molecules cross the membrane through diffusion facilitated by transport proteins. Channel diffusion and facilitated diffusion are forms of passive transport, as the molecules diffuse under their own energy down their concentration gradients. In secondary active transport, the energy released from the diffusion of one set of molecules (the driving substrate) down its concentration gradient powers the transport of another set (the driven substrate) either down or up its own gradient. In primary active transport, energy released from cellular metabolism powers the transport of molecules up their concentration gradients.

porter must be able to undergo a conformational change when bound to neither substrate, so as to complete the cycle and return the binding sites to the side from which driving and driven substrates both move.

Sodium ions are usually the driving substrates in the co-transport systems of animal cells, which maintain high concentrations of these ions through primary active transport. The driven substrates include a variety of sugars, amino acids, and other ions. During the absorption of nutrients, for example, sugars and amino acids are removed from the intestine by co-transport with sodium ions. After passing across the glomerular filter in the kidney, these substrates are returned to the body by the same system. Plant and bacterial cells usually use hydrogen ions as the driving substrate; sugars and amino acids are the most common driven substrates. When the bacterium *Escherichia coli* must metabolize lactose, it co-transports hydrogen ions with lactose (which can reach a concentration 1,000 times higher than that outside the cell).

Primary active transport. *The sodium pump.* Human red blood cells contain a high concentration of potassium and a low concentration of sodium, yet the plasma bathing the cells is high in sodium and low in potassium. When whole blood is stored cold under laboratory conditions, the cells lose potassium and gain sodium until the concentrations across the membrane for both ions are at equilibrium. When the cells are restored to body temperature and given appropriate nutrition, they extrude sodium and take up potassium, transporting both ions against their respective gradients until the previous high concentrations are reached. This ion pumping is linked directly to the hydrolysis of adenosine triphosphate (ATP), the cell's repository of metabolic energy (see above *The nature and function of cells: Coupled chemical reactions*). For every molecule of ATP split, three ions of sodium are pumped out of the cell and two of potassium are pumped in.

Active transport and ATP hydrolysis

An enzyme called sodium–potassium-activated ATPase has been shown to be the sodium pump, the protein that transports the ions across the cell membrane while splitting ATP. Widely distributed in the animal kingdom and always associated with the plasma membrane, ATPase is found at high concentration in cells that pump large amounts of sodium (*e.g.,* in mammalian kidneys, in salt-secreting glands of marine birds, and in the electric organs of eels). The enzyme, an intrinsic protein, exists in two major conformations whose interconversion is driven by the splitting of ATP or by changes in the transmembrane flows of sodium and potassium. When only sodium is present in the cell, the inorganic phosphate split from ATP during hydrolysis is transferred to the enzyme. Release of the chemically bound phosphate from the enzyme is catalyzed by potassium. Thus, the complete action of ATP splitting has been demonstrated to require both sodium (to catalyze the transfer of the phosphate to the enzyme) and potassium (to catalyze the release of the phosphate and free the enzyme for a further cycle of ATP splitting). Apparently, only after sodium has catalyzed the transferal of the phosphate to the enzyme can it be transported from the cell. Similarly, only after potassium has released the phosphate from the enzyme can it be transported into the cell. This overall reaction, completing the cycle of conformational changes in the enzyme, involves a strict coupling of the splitting of ATP with the pumping of sodium and potassium. It is this coupling that creates primary active transport.

The sodium pump extrudes one net positive charge during each cycle of ATP splitting. This flow of current induces an electric potential across the membrane that adds to the potentials brought about by the diffusion of ions through gated channels. The pump's contribution to the overall potential is important in certain specialized nerve cells.

Calcium pumps. Many animal cells can perform a primary active transport of calcium out of the cell, developing a 10,000-fold gradient of that ion. Calcium-activated ATPases have been isolated and shown to be intrinsic proteins straddling the membrane and undergoing conformational changes similar to those of the sodium–potassium-activated ATPase. When a rise in the concentration of cellular calcium results from the opening of calcium-selective channels, the membrane's calcium pumps restore the low concentration.

Hydrogen ion pumps. Hydrochloric acid is produced in the stomach by the active transport of hydrogen ions from the blood across the stomach lining, or gastric mucosa. Hydrogen concentration gradients of nearly 1,000,000 can be achieved by a hydrogen–potassium-activated ATP-splitting intrinsic protein in the cells lining the stomach. Apart from its specific ion requirements, the properties of this enzyme are remarkably similar to those of the sodium–potassium-activated enzyme and the calcium-activated enzyme. Other hydrogen-pumping, ATP-splitting primary active transporters occur in intracellular organelles, in bacteria, and in plant cells (see below *The mitochondrion and the chloroplast*). The steep gradient of hydrogen ions represents a store of energy that can be harnessed to the accumulation of nutrients or, in the case of bacterial flagella, to the powering of cell movement.

Transport of particles. Molecules of the size of proteins or nucleic acids cannot cross the plasma membrane by simple diffusion or by any of the transport routes considered above. Yet it is certain that such large molecules enter and leave the cell. In bringing about transmembrane movements of large molecules, the plasma membrane itself undergoes concerted movements during which part of the fluid medium outside of the cell is internalized (endocytosis) or part of the cell's internal medium is externalized (exocytosis). The details of these movements are only partly understood, but they seem to involve a fusion between membrane surfaces, followed by the re-formation of intact membranes.

Endocytosis. In this process the plasma membrane engulfs portions of the external medium, forms an almost complete sphere around it, and then draws the membrane-bounded vesicle into the cell. Several types of endocytosis have been distinguished: in pinocytosis, the vesicles are small and contain fluid; in phagocytosis, the vesicles are larger and contain solid matter; in receptor-mediated endocytosis, material binds to a specific receptor on the external face of the plasma membrane, triggering the process by which it is engulfed. Cholesterol enters cells by the last route.

Exocytosis. In this process material synthesized within the cell, and packaged into membrane-bounded vesicles, is exported from the cell following the fusion of the vesicles with the external plasma membrane. The materials so exported are cell-specific protein products, neurotransmitters, and a variety of other molecules. (W.D.S.)

Internal membranes

The presence of internal membranes distinguishes eukaryotic cells (cells with a nucleus) from prokaryotic cells (those without a nucleus). Prokaryotic cells are small (one to five micrometres in length) and contain only a single plasma membrane; metabolic functions are often confined to different patches of the membrane rather than to areas in the body of the cell. Typical eukaryotic cells, by contrast, are much larger, the plasma membrane constituting only 10 percent or less of the total cellular membrane. Metabolic functions in these cells are carried out in the organelles, compartments sequestered from the cell body, or cytoplasm, by internal membranes.

This section discusses internal membranes as structural and functional components in the organelles and vesicles of eukaryotic cells. The principal organelles—the nucleus, mitochondrion, and (in plants) chloroplast—are discussed elsewhere (see below *The nucleus* and *The mitochondrion and the chloroplast*). Of the remaining organelles, the lysosomes, peroxisomes, and (in plants) glyoxysomes enclose extremely reactive by-products and enzymes. Internal membranes form the mazelike endoplasmic reticulum, where cell membrane proteins and lipids are synthesized, and they also form the stacks of flattened sacs called the Golgi apparatus, which is associated with the transport and modification of lipids, proteins, and carbohydrates. Finally, internal cell membranes can form storage and transport vesicles and the vacuoles of plant cells. Each membrane structure has its own distinct composition of proteins and lipids enabling it to carry out unique functions.

GENERAL FUNCTIONS AND CHARACTERISTICS

Like the plasma membrane, membranes of some organelles contain transport proteins, or permeases, that allow chemical communication between organelles. Permeases in the lysosomal membrane, for example, allow amino acids generated inside the lysosome to cross into the cytoplasm, where they can be used for the synthesis of new proteins. Communication between organelles is also achieved by the membrane budding processes of endocytosis and exocytosis, which are essentially the same as in the plasma membrane (see above *The plasma membrane: Transport across the membrane*). On the other hand, the biosynthetic and degradative processes taking place in different organelles may require conditions greatly different from those of other organelles or of the cytosol (the fluid part of the cell surrounding the organelles). Internal membranes maintain these different conditions by isolating them from one another. For example, the internal space of lysosomes is much more acidic than that of the cytosol—pH 5 as opposed to pH 7—and is maintained by specific proton-pumping transport proteins in the lysosome membrane.

Separating incompatible environments and metabolic processes

Another function of organelles is to prevent competing enzymatic reactions from interfering with one another. For instance, essential proteins are synthesized on the rough endoplasmic reticulum and in the cytosol, while unwanted proteins are broken down in the lysosomes and also, to some extent, in the cytosol. Similarly, fatty acids are made in the cytosol and then either broken down in the mitochondria for the synthesis of ATP or degraded in the peroxisomes with concomitant generation of heat. These processes must be kept isolated. Organelle membranes also prevent potentially lethal by-products or enzymes from attacking sensitive molecules in other regions of the cell by sequestering such degradative activities in their respective membrane-bounded compartments.

The internal membranes of eukaryotic cells differ both structurally and chemically from the plasma membrane. Like the plasma membrane, they are constructed of a phospholipid bilayer into which are embedded, or bound, specific membrane proteins (see above *The plasma membrane: Chemical composition and structure of the membrane*). The three major lipids forming the plasma membrane—phospholipids, cholesterol, and glycolipids—are also found in the internal membranes, but in different concentrations. Phospholipid is the primary lipid forming all cellular membranes. Cholesterol, which contributes to the fluidity and stability of all membranes, is found in internal membranes at about 25 percent of the concentration in the plasma membrane. Glycolipids are found only as trace components of internal membranes, whereas they constitute approximately 5 percent of the plasma membrane lipid.

CELLULAR ORGANELLES AND THEIR MEMBRANES

The vacuole. Most plant cells contain one or more membrane-bound vesicles called vacuoles. Within the vacuole is the cell sap, a water solution of salts and sugars kept at high concentration by the active transport of ions through permeases in the vacuole membrane. This high concentration causes the entry, via osmosis, of water into the vacuole, which in turn expands the vacuole and generates a hydrostatic pressure, called turgor, that presses the cell membrane against the cell wall. Turgor is the cause of rigidity in living plant tissue.

In the mature plant cell, as much as 90 percent of cell volume may be taken up by a single vacuole; immature cells typically contain several smaller vacuoles.

The lysosome. Potentially dangerous hydrolytic enzymes functioning in acidic conditions (pH 5) are segregated in the lysosomes to protect the other components of the cell from random destruction. Lysosomes are bound by a single phospholipid bilayer membrane. They vary in size and probably originate by budding from the Golgi membranes. Enzymes known to be present in the lysosomes include hydrolases, which degrade proteins, nucleic acids, lipids, glycolipids, and glycoproteins; they are most active in the acidity maintained in the lysosomes. Lysosomes fuse with vacuoles containing material from inside or outside the cell to be digested. After the material is broken down, lipids and amino acids are transported across the lysosomal membrane by permeases for use in biosynthesis. The remaining debris generally remains within the lysosome and is called a residual body.

Microbodies. Microbodies are roughly spherical in shape, bound by a single membrane, and are usually 0.5 to one micrometre in diameter. There are several types, by far the most common of which is the peroxisome. Peroxisomes derive their name from hydrogen peroxide, a reactive intermediate in the process of molecular breakdown that occurs in the microbody. Peroxisomes contain type II oxidases, which are enzymes that use molecular oxygen in reactions to oxidize organic molecules. A product of these reactions is hydrogen peroxide, which is further metabolized into water and oxygen by the enzyme catalase, a predominant constituent of peroxisomes. In addition, peroxisomes contain other enzyme systems that degrade various lipids.

The peroxisome

The plant glyoxysome is a peroxisome that also contains the enzymes of the glyoxylate cycle, which is crucial to the conversion of fat into carbohydrate.

The endoplasmic reticulum. The endoplasmic reticulum (ER) is a system of membranous vesicles extending throughout the cytoplasm. Often, it constitutes more than half of the total membrane in the cell. This structure was first noted in the late 19th century, when studies of stained cells indicated the presence of some type of extensive cytoplasmic structure, then termed the gastroplasm. The electron microscope made possible the study of the morphology of this organelle in the 1940s, when it was given its present name.

The endoplasmic reticulum can be classified in two functionally distinct forms, the smooth endoplasmic reticulum (SER) and the rough endoplasmic reticulum (RER). The morphological distinction between the two is the presence of protein-synthesizing particles, called ribosomes, attached to the outer surface of the RER.

The smooth endoplasmic reticulum. The functions of the SER, a meshwork of fine tubular membrane vesicles, vary considerably from cell to cell. One important role is the synthesis of phospholipids and cholesterol, which are major components of the plasma and internal membranes. Phospholipids are formed from fatty acids, glycerol phosphate, and other small, water-soluble molecules by enzymes bound to the ER membrane with their active sites facing the cytosol. Some phospholipids remain in the ER membrane, where, catalyzed by specific enzymes within the membranes, they can "flip" from the cytoplasmic side of the bilayer, where they were formed, to the exoplasmic, or inner, side. This process ensures the symmetrical growth of the ER membrane. Other phospholipids are transferred through the cytoplasm to other membranous structures, such as the plasma membrane and the mitochondrion, by special phospholipid transfer proteins.

In liver cells, the SER is specialized for the detoxification of a wide variety of compounds produced by metabolic processes. Liver SER contains a number of enzymes called cytochrome P450, which catalyze the breakdown of carcinogens and other organic molecules. In cells of the adrenal glands and gonads, cholesterol is modified in the SER at one stage of its conversion to steroid hormones. Finally, the SER in muscle cells, known as the sarcoplasmic reticulum, sequesters calcium ions from the cytoplasm. When the muscle is triggered by nerve stimuli, the calcium ions are released, causing muscle contraction.

The rough endoplasmic reticulum. The RER is generally a series of connected vesicles and flattened sacs. It plays a central role in the synthesis and export of proteins and glycoproteins and is best studied in the secretory cells specialized in these functions. The many secretory cells in the body include liver cells secreting serum proteins such as albumin, endocrine cells secreting peptide hormones such as insulin, salivary gland and pancreatic acinar cells secreting digestive enzymes, mammary gland cells secreting milk proteins, and cartilage cells secreting collagen and proteoglycans.

Production of proteins in the ribosomes

Ribosomes are particles that synthesize proteins from amino acids. They are composed of four RNA molecules and about 50 proteins assembled into a large and a small subunit. Ribosomes are either free (*i.e.,* not bound to membranes) in the cytoplasm of the cell or bound to the RER. Lysosomal enzymes, proteins destined for the ER, Golgi, and plasma membranes, and proteins to be secreted from the cell are among those synthesized on membrane-bound ribosomes. Fabricated on free ribosomes are proteins remaining in the cytosol, those bound to the internal surface of the plasma membrane, as well as those to be incorporated into the nucleus, mitochondria, chloroplasts, peroxisomes, and other organelles. In order to segregate their sites of synthesis, the cell must have a mechanism that can differentiate between these two classes of proteins. This mechanism cannot involve the ribosomes themselves, since those ribosomes bound to the ER membrane are no different from those found in the cytoplasm. In 1971 the biochemists Gunter Blobel and David Sabatini suggested that the amino-terminal portion of the protein (the first part of the molecule to be made) could act as a "signal sequence," which would facilitate attachment of the growing protein to the ER membrane and lead it either into the membrane or through the membrane into the ER lumen, or interior.

The signal hypothesis has been substantiated by a large body of experimental evidence. It is now known that "translation" of the blueprint for a specific protein encoded in a messenger RNA molecule begins on a free ribosome. As the growing protein, with the signal sequence at its amino-terminal end, emerges from the ribosome, the sequence binds to a complex of six proteins and one RNA molecule known as the signal recognition particle (SRP). The SRP also binds to the ribosome to halt further formation of the protein. The membrane of the ER contains receptor sites that bind the SRP-ribosome complex to the RER membrane. Upon binding, translation resumes, with the SRP dissociating from the complex and the signal sequence and remainder of the nascent protein threading through the membrane into the ER lumen. At that point, the protein is permanently segregated from the cytosol. In most cases, the signal sequence is cleaved from the protein by an enzyme called signal peptidase as it emerges on the luminal surface of the ER membrane. In addition, in a process known as glycosylation, oligosaccharide (complex sugar) chains are often added to the protein to form a glycoprotein. Inside the ER lumen, the protein folds into its characteristic three-dimensional conformation.

Transport of proteins from endoplasmic reticulum to Golgi apparatus

Within the lumen, proteins that will be secreted from the cell diffuse into the transitional portion of the ER, a region that is largely free of ribosomes. There, the molecules are packaged into small, membrane-bounded transport vesicles, which separate from the ER membrane and move through the cytoplasm to a target membrane, usually the Golgi complex. There the transport vesicle membrane fuses with the Golgi membrane and the contents of the vesicle are delivered into the lumen of the Golgi. This, like all processes of vesicle budding and fusion, preserves the sidedness of the membranes; that is, the cytoplasmic surface of the membrane always faces outward, and the luminal contents are always sequestered from the cytoplasm.

Certain nonsecretory proteins made on the RER remain part of the membrane system of the cell. These membrane proteins have, in addition to the signal sequence, one or more anchor regions composed of lipid-soluble amino acids. The amino acids prevent passage of the protein completely into the ER lumen by anchoring it into the phospholipid bilayer of the ER membrane.

The Golgi apparatus. The Golgi complex is the site of the modification, completion, and export of secretory proteins and glycoproteins. This organelle, first described by the Italian cytologist Camillo Golgi in 1898, has a characteristic structure composed of five to eight flattened, disk-shaped, membrane-defined cisternae, or vesicles, arranged in a stack. Though the mechanisms are not well understood, it is thought that the Golgi is the principal director of protein traffic in the cell. Secretory proteins and glycoproteins, plasma membrane proteins and glycoproteins, lysosomal proteins, and some glycolipids all pass through the Golgi structure at some point in their maturation. In plant cells, much of the cell-wall material passes through the Golgi as well.

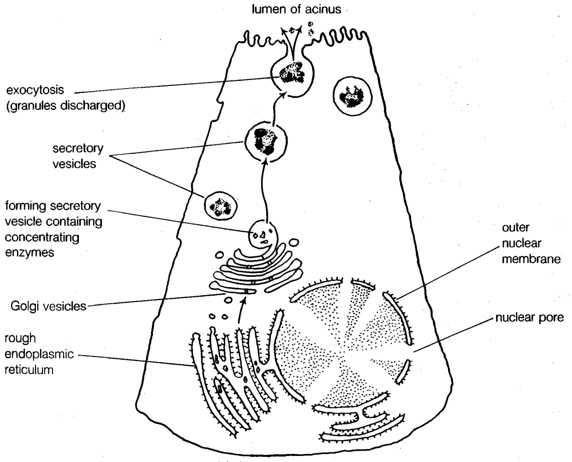

Figure 6: *The path taken by newly synthesized secretory proteins in a rat pancreatic acinar cell.*
Immediately after synthesis, the secretory proteins are found in the lumen of the rough endoplasmic reticulum. Transfer vesicles transport them to the Golgi complex. Next they are concentrated in secretory vesicles. The secretory vesicles are formed under the region of the plasma membrane that faces a ductule of the acinus. Exocytosis of the vesicles releases the contents into the ductule; from there, the contents move to the intestine, where they become digestive enzymes.

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The Golgi apparatus itself is structurally polarized, consisting of a "cis" face near the transitional region of the RER, a medial segment, and a "trans" face near the plasma membrane. These faces are biochemically distinct, and the enzymatic content of each segment is markedly different. The cis face membranes are generally thinner than the others.

As the secretory proteins move through the Golgi, a number of chemical modifications may transpire. Important among these is the modification of carbohydrate groups. As described above, many secretory proteins are glycosylated in the ER. In the Golgi, specific enzymes modify the oligosaccharide chains of the glycoproteins by removing certain mannose residues and adding other sugars, such as galactose and sialic acid. These enzymes are known collectively as glycosidases and glycosyltransferases. The role of the oligosaccharide chains is not known. Some secretory proteins will cease to be transported if their carbohydrate groups are modified incorrectly or not permitted to form. In some cases the carbohydrate groups are necessary for the stability or activity of the protein or for targeting the molecule for a specific destination.

Also within the Golgi or secretory vesicles are proteases that cut many secretory proteins at specific amino acid positions. This often results in activation of the secretory protein, an example being the conversion of inactive proinsulin to active insulin by removing a series of amino acids.

Discharge of proteins from the cell

Secretory vesicles. Proteins move from the cis to the trans face of the Golgi in transport vesicles that bud and fuse with other membranes much in the same manner as the transport from the ER to Golgi. Often, this trip across the Golgi acts to concentrate the secretory product. At the trans face, the product may move into one specific area and be enclosed in a vacuole. The vacuole may then become a secretory vesicle that moves to the plasma membrane, where the contents can be discharged from the cell. This is a closely regulated process; discharge takes place only at the appropriate membrane to avoid the dangerous or wasteful discharge of products. For example, pancreatic enzymes are secreted from the acinar cells into ductules, which carry the enzymes not into the blood but into the intestine.

The stimulus for discharge is most often a nervous or hormonal signal. Discharge itself is generally triggered by a rise in the concentration of calcium ions in the cell.

The action is one of exocytosis: the vesicle and the plasma membrane fuse, forming a channel by which the proteins and glycoproteins in the vesicle can be released.

As secretory vesicles fuse with the plasma membrane, the area of the plasma membrane increases. Normal size is regained by the recycling of membrane components through endocytosis. Regions bud from the plasma membrane and then fuse with the membrane of the Golgi or lysosomes.

SORTING OF PRODUCTS BY CHEMICAL RECEPTORS

Not all proteins synthesized on the ER are destined for export. Many, such as the hydrolases in lysosomes, remain inside the cell; others become anchored in the membrane of internal organelles or in the plasma membrane. It is presumed that each protein has some type of marker that fits a specific location in the cell.

Proteins synthesized on free ribosomes may have segments that bind to specific receptors on the outer membrane of mitochondria, chloroplasts, or peroxisomes, allowing these proteins to be taken up only by these organelles. In the case of proteins synthesized on the RER, both the hydrolases destined for lysosomes and the secretory proteins are found initially in the same portion of the ER lumen. Studies have shown that these can be distinguished on the basis of their carbohydrate residues. The carbohydrate residues of lysosomal enzymes become modified in the cis-Golgi by the addition of certain phosphate groups. This critical modification allows the enzymes to bind to specific receptors on the membrane of the Golgi, which then directs them into vesicles leading to a lysosome rather than a secretory vesicle. In the lysosomes, proton pumps create an acidic environment that causes the release of the lysosomal enzyme from the membrane-bound receptors.

It is now thought that much of this sorting activity is mediated by coated vesicles containing the same fibrous outer protein, clathrin, used in endocytosis. These sorting vesicles also contain associated smaller proteins. Some of the smaller proteins may act as receptors for the transported proteins, while others may be markers on the surface of vesicles, allowing the proper receptors on the appropriate target membrane to recognize and accept the coated vesicle and its contents. (H.F.L./Ch.C./M.Cu.)

The nucleus

The nucleus is the information centre of the cell in all higher organisms. It is separated from the cytoplasm by the nuclear envelope, and it houses the double-stranded, spiral-shaped deoxyribonucleic acid (DNA) molecules, which contain the genetic information necessary for the cell to retain its unique character as it grows and divides.

Distinction between prokaryotic and eukaryotic cells

The presence of a nucleus distinguishes the eukaryotic cells of multicellular organisms from the prokaryotic, one-celled organisms, such as bacteria. In contrast to the higher organisms, bacteria do not have nuclei, so that their DNA is maintained in the same compartment as the other cellular components.

The primary function of the nucleus is the expression of selected subsets of the genetic information encoded in the DNA double helix. Each subset of a DNA chain, called a gene, codes for the construction of a specific protein out of a chain of amino acids. Information in DNA is not decoded directly into proteins, however. First it is transcribed, or copied, into a range of messenger ribonucleic acid (mRNA) molecules, each of which encodes the information for one protein (or more than one protein in bacteria). The mRNA molecules are then transported through the nuclear envelope into the cytoplasm, where they are translated, serving as templates for the synthesis of specific proteins.

The nucleus must not only synthesize the mRNA for many thousands of proteins, but it must also regulate the amounts synthesized and supplied to the cytoplasm. Furthermore, the amounts of each type of mRNA supplied to the cytoplasm must be regulated differently in each type of cell. In addition to mRNA, the nucleus synthesizes and exports other classes of RNA involved in the mechanisms of protein synthesis.

STRUCTURAL ORGANIZATION OF THE NUCLEUS

DNA packaging. The nucleus of the average human cell is only six micrometres (6×10^{-6} metre) in diameter, yet it contains about 1.8 metres of DNA. This is distributed among 46 chromosomes, each consisting of a single DNA molecule about 40 millimetres (1½ inches) long. The extraordinary packaging problem this poses can be envisaged by a scale model enlarged a million times. On this scale a DNA molecule would be a thin string two millimetres thick, and the average chromosome would contain 40 kilometres (25 miles) of DNA. With a diameter of only six metres, the nucleus would contain 1,800 kilometres of DNA.

These contents must be organized in such a way that they can be copied into RNA accurately and selectively. DNA is not simply crammed or wound into the nucleus like a ball of string; rather, it is organized, by molecular interaction with specific nuclear proteins, into a precisely packaged structure. This combination of DNA with proteins creates a dense, compact fibre called chromatin. An extreme example of the ordered folding and compaction that chromatin can undergo is seen during cell division, when the chromatin of each chromosome condenses and is divided between two daughter cells (see below *Cell division and growth*).

Histone core of the nucleosome

Nucleosomes: the subunits of chromatin. The compaction of DNA is achieved by winding it around a series of small proteins called histones. Histones are composed of positively charged amino acids that bind tightly to and neutralize the negative charges of DNA. There are five classes of histone. Four of them, called H2A, H2B, H3, and H4, contribute two molecules each to form an octamer, an eight-part core around which two turns of DNA are wrapped. The resulting beadlike structure is called the nucleosome. The DNA enters and leaves a series of nucleosomes, linking them like beads along a string in lengths that vary between species of organism or even between different types of cell within a species. A string of nucleosomes is then coiled into a solenoid configuration by the fifth histone, called H1. One molecule of H1 binds to the site at which DNA enters and leaves each nucleosome, and a chain of H1 molecules coils the string of nucleosomes into the solenoid structure of the chromatin fibre (see Figure 7).

After B. Alberts *et al.*, *Molecular Biology of The Cell* (1983); Garland Publishing, Inc.

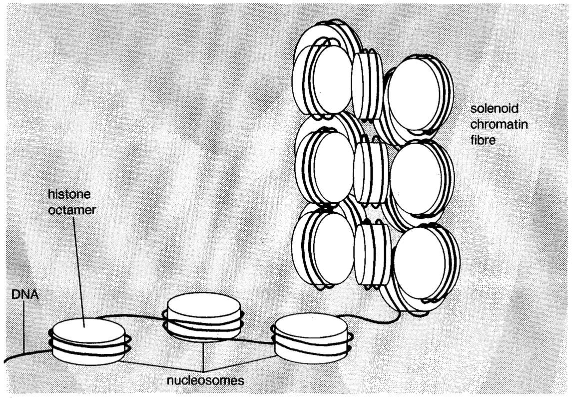

Figure 7: DNA wrapped around clusters of histone proteins to form nucleosomes, which can coil to form solenoids.

Nucleosomes not only neutralize the charges of DNA, but they have other consequences. First, they are an efficient means of packaging. DNA becomes compacted by a factor of six when wound into nucleosomes and by a factor of about 40 when the nucleosomes are coiled into a solenoid chromatin fibre. The winding into nucleosomes also allows some inactive DNA to be folded away in inaccessible conformations, a process that probably contributes to the selectivity of gene expression.

Organization of chromatin fibre. Several studies indicate that chromatin is organized into a series of large radial loops anchored to specific scaffold proteins. Each loop consists of a chain of nucleosomes and may be related to units of genetic organization. This radial arrangement of chromatin loops compacts DNA about a thousandfold. Further compaction is achieved by a coiling of the entire looped chromatin fibre into a dense structure called a chromatid, two of which form the chromosome. (The bands that appear in fluorescent micrographs of chromosomes may correspond to coils of looped chromatin fibre.) During cell division, this coiling produces a 10,000-fold compaction of DNA. After division, the condensed chromosomes partly unfold as a nuclear envelope reforms around them. However, the organization of most DNA into nucleosomes persists within the active nucleus.

The nuclear envelope. The nuclear envelope is a double membrane composed of an outer and an inner phospholipid bilayer. The thin space between the two layers appears to connect with the lumen of the rough endoplasmic reticulum (RER), and the outer layer is an extension of the outer face of the RER.

The nuclear pore

The inner surface of the nuclear envelope has a protein lining called the nuclear lamina, which binds to chromatin and other contents of the nucleus. The entire envelope is perforated by numerous nuclear pores. These transport routes are fully permeable to small molecules up to the size of the smallest proteins, but they form a selective barrier against movement of larger molecules. Each pore is surrounded by an elaborate protein structure called the nuclear pore complex, which probably selects large molecules for entrance into the nucleus. Entering the nucleus through the pores are the nucleotide building blocks of DNA and RNA, as well as adenosine triphosphate, which provides the energy for synthesizing genetic material. Histones and other large proteins must also pass through the pores. These molecules probably have structures on their surface that signal admittance by the nuclear pore complexes. The complexes also probably regulate the export from the nucleus of RNA and subunits of ribosomes.

DNA in prokaryotes is also organized in loops and is bound to small proteins resembling histones, but this structure is not enclosed by a nuclear membrane.

GENETIC ORGANIZATION OF THE NUCLEUS

The structure of DNA. Several features are common to the genetic structure of most organisms. First is the double-stranded DNA. Each strand of this molecule is a series of nucleotides, and each nucleotide is composed of a sugar–phosphate compound attached to one of four nitrogen-containing bases. The sugar–phosphate compounds link together to form the backbone of the strand. Each of the bases strung along the backbone is chemically attracted to a corresponding base on the parallel strand of the DNA molecule. This base pairing joins the two strands of the molecule much as rungs join the two sides of a ladder, and the chemical bonding of the base pairs twists the doubled strands into a spiral, or helical, shape.

The four nucleotide bases are adenine, cytosine, guanine, and thymine. DNA is composed of millions of these bases strung in an almost limitless variety of sequences. It is in the sequence of bases that the genetic information is contained, each sequence determining the sequence of amino acids to be connected into proteins. A nucleotide sequence sufficient to encode one protein is called a gene. Genes are interspersed along the DNA molecule with other sequences that do not encode proteins. Some of these so-called untranslated regions regulate the activity of the adjacent genes, for example, by marking the points at which enzymes begin and cease transcribing DNA into RNA (see below *Genetic expression through RNA*).

A complete set of DNA in all cells of the body

Rearrangement and modification of DNA. Different cells of the body owe their specialized structures and functions to different genes. This does not mean that the set of genetic information varies among the cells of the body. Indeed, for each cell the entire DNA content of the chromosomes is usually duplicated exactly from generation to generation, and in general the genetic content and arrangement is strikingly similar among different cell types of the same organism. The clearest evidence of this comes from several species of plants in which it is possible to grow entire new plants from single cells, indicating that the cells contain all of the genetic information required for the plant's entire life cycle. The closest equivalent experiment in animals is the hatching of tadpoles from frogs' eggs whose nuclei have been replaced by nuclei of other cells from an adult frog. When cells from the intestine of a frog embryo are used as donors of nuclei, it is possible to raise fertile adult frogs carrying genetic markers from the donor intestine cell and none from the recipient egg. These experiments stop short of showing that all the cells of an organism contain copies of all genes, but they do establish that the nuclei of some highly differentiated cells retain functional copies of sufficient genes to specify, in these cases, a complete, swimming tadpole or a fertile adult frog.

Differentiation of cells, therefore, can occur without the loss or irreversible inactivation of unnecessary genes. This conclusion is reinforced by studies showing the presence of specific genes in a range of adult tissues. For example, normal copies of the genes encoding hemoglobin are present in the same numbers in red blood cells, which make hemoglobin, as in a range of other types of cell, which do not.

Despite the general uniformity of genetic content in all the cells of an organism, studies have shown a few clear examples in some organisms of programmed, reversible change in the DNA of developing tissues. One of the most dramatic rearrangements of DNA occurs in the immune systems of mammals. The body's defense against invasion by foreign organisms involves the synthesis of a vast range of antibodies by lymphocytes (a type of white blood cell). Antibodies are proteins that bind to specific invading molecules or organisms and either inactivate them or signal their destruction. The binding sites on each antibody molecule are formed by one light and one heavy amino acid chain, which are encoded by different segments of the DNA in the lymphocyte nucleus. Upon stimulation by an invading particle, these DNA segments undergo slight rearrangements, resulting in the synthesis of a great variety of antibodies.

Trypanosome parasites, which cause sleeping sickness, undergo DNA rearrangement in order to overcome the versatility of their hosts' antibody production. The parasites are covered by a thick coat of glycoprotein (a protein with sugars attached). Given time, host organisms can overcome infection by producing antibodies to the parasites' glycoprotein coat, but this reaction is anticipated and evaded by the selective rearrangement of the trypanosomes' DNA encoding the glycoprotein, thus constantly changing the surface presented to the hosts' immune system.

Inactivation of genes

Careful comparisons of gene structure have also revealed a structural modification of DNA that occurs mostly in inactive genes. This modification involves the addition of a methyl group to cytosine bases in the DNA chain. It appears to have the role in vertebrate animals of inactivating genes that are not required for expression in a particular type of cell. An important feature of the methylation of cytosine lies in its ability to be copied, so that methyl groups in a dividing cell's DNA will result in methyl groups in the same positions in the DNA of both daughter cells.

From studies comparing the structure of DNA in different cells of the same organism there emerges a picture of remarkable consistency. Modifications and rearrangements of the nucleotide sequences in DNA are more exceptions than the rule, although they do cause significant changes in the structure and function of some cells. But it is in the transcription of the genetic code from DNA to RNA, and the translation of that code from RNA into protein, that the greatest modulation of genetic information occurs. This process, called genetic expression, takes place over several stages, and at each stage is the potential for further differentiation of cell types.

GENETIC EXPRESSION THROUGH RNA

As explained above, genetic information is encoded in the sequences of the four nucleotide bases making up a DNA molecule. One of the two DNA strands is transcribed exactly into RNA, with the exception that the thymine base of DNA is replaced by uracil. RNA also contains a slightly different sugar component (ribose) from that of DNA (deoxyribose) in its connecting sugar–phosphate chain. Unlike DNA, which is stable throughout the cell's life and individual strands of which are even passed on to many cell generations, RNA is unstable. It is continuously broken down and replaced, enabling the cell to change its patterns of protein synthesis.

Apart from mRNA, which encodes proteins, other classes of RNA are made by the nucleus. These include ribosomal RNA (rRNA), which forms part of the ribosomes and is also exported to the cytoplasm to help translate the information in mRNA into proteins. Ribosomal RNA is synthesized in a specialized region of the nucleus called the nucleolus, which appears as a dense area within the nucleus and contains the genes that encode rRNA (see Figure 8). This is also the site of assembly of ribosome subunits from rRNA and ribosomal proteins. Ribosomal proteins are synthesized in the cytoplasm and transported to the nucleus for subassembly in the nucleolus. The subunits are then returned to the cytoplasm for completion. Another class of RNA synthesized in the nucleus is transfer RNA (tRNA), which serves as an adaptor matching individual amino acids to the nucleotide triplets of mRNA during protein synthesis.

By courtesy of E.G. Jordan and J. McGovern

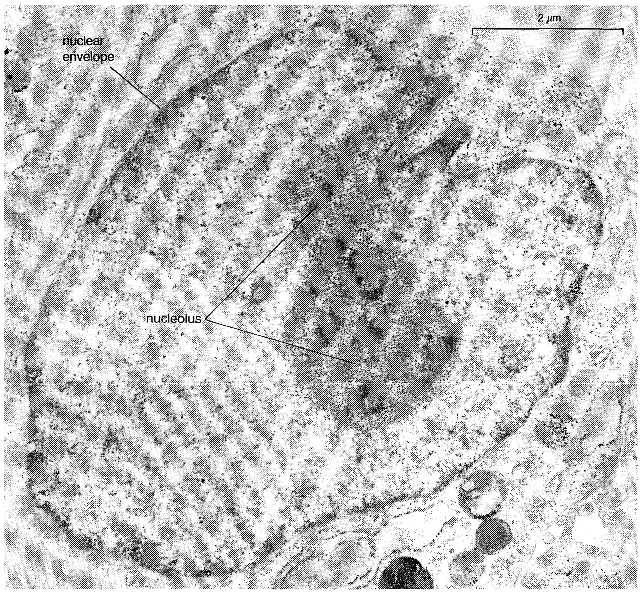

Figure 8: Electron micrograph of a thin section of the nucleus of a human cell, showing the nuclear envelope and nucleolus.

Three main RNA-synthesizing enzymes

RNA synthesis. The synthesis of RNA is performed by enzymes called RNA polymerases. There are three main RNA polymerases in higher organisms, designated I, II, and III (or sometimes A, B, and C). Each is a complex protein consisting of many subunits. RNA polymerase I synthesizes three of the four types of rRNA (called 18S, 28S, and 5.8S RNA); therefore it is active in the nucleolus, where the genes encoding these rRNA molecules reside. RNA polymerase II synthesizes mRNA, though its initial products are not mature RNA but larger precursors, called heterogeneous nuclear RNA, which are completed later (see below *Processing of mRNA*). The products of RNA polymerase III include tRNA and the fourth RNA component of the ribosome, which is called 5S RNA.

All three polymerases start RNA synthesis at specific sites on DNA and proceed along the molecule, linking selected nucleotides sequentially until they come to the end of the gene and, therefore, the end of the growing RNA chain. Energy for RNA synthesis comes from high-energy phosphate linkages contained in the nucleotide precursors of RNA. Each unit of the final RNA product is essentially a sugar, a base, and one phosphate, but the building material consists of a sugar, a base, and three phosphates. During synthesis two phosphates are cleaved and discarded for each nucleotide incorporated into RNA, and the energy released from the phosphate bonds is used to link the nucleotides. The crucial feature of RNA synthesis is that the sequence of nucleotides joined into a growing RNA chain is specified by the sequence of nucleotides in the DNA template: each adenine in DNA specifies uracil in RNA, each cytosine specifies guanine, each guanine specifies cytosine, and each thymine in DNA specifies adenine. In this way the information encoded in each gene is transcribed into RNA for translation by the protein-synthesizing machinery of the cytoplasm.

In addition to specifying the sequence of amino acids to be polymerized into proteins, the nucleotide sequence of DNA contains supplementary information. For example, short sequences of nucleotides determine the initiation site for each RNA polymerase, specifying where and when RNA synthesis should occur. In the case of RNA polymerases I and II, the sequences specifying initiation sites lie just ahead of the genes. In contrast, the equivalent information for RNA polymerase III lies within the gene, that is, within the region of DNA to be copied into RNA. The initiation site on a segment of DNA is called a promoter. The promoters of different genes have some nucleotide sequences in common, but they differ in others. The differences in sequence are recognized by specific proteins called transcription factors, which are necessary for the expression of particular types of genes and which contribute to differences in the gene expression of different types of cells.

Processing of mRNA. During and after synthesis, mRNA precursors undergo a complex series of changes before the mature molecules are released from the nucleus. First, a modified nucleotide is added to the start of the RNA molecule by a reaction called capping. This cap later binds to a ribosome in the cytoplasm. The synthesis of mRNA is not terminated simply by the RNA polymerase's detachment from DNA, as usually occurs, but by chemical cleavage of the RNA chain. Many (but not all) types of mRNA have a simple polymer of adenosine residues added to their cleaved ends.

In addition to these modifications of the termini, startling discoveries in 1977 revealed that portions of newly synthesized RNA molecules are cut out and discarded. In many genes, the regions coding for proteins are interrupted by intervening sequences of nucleotides called introns. These introns must be excised from the RNA copy before it can be released from the nucleus as a functional mRNA. The number and size of introns within a gene vary greatly, from no introns at all to more than 50. The sum of the lengths of these intervening sequences is sometimes longer than the sum of the regions coding for proteins.

RNA splicing

The removal of introns, called RNA splicing, appears to be mediated by small nuclear ribonucleoprotein particles (snRNP's). These particles have RNA sequences that are complementary to the junctions between introns and adjacent coding regions. By binding to the junction ends, an snRNP twists the intron into a loop. It then excises the loop and splices the coding regions.

REGULATION OF GENETIC EXPRESSION

Although all the cell nuclei of an organism generally carry the same genes, there are conspicuous differences between the specialized cell types of the body. The source of these differences lies not so much in the occasional modification of DNA, as outlined above, but in the selective expression of DNA through RNA; in particular, it can be traced to processes regulating the amounts and activities of mRNA both during and after its synthesis in the nucleus.

Regulation of RNA synthesis. The first level of regulation is mediated by variations in chromatin structure. In order to be transcribed, a gene must be assembled into a structurally distinct form of active chromatin. Although this is poorly characterized, it appears to correspond to the partial unfolding of chromatin. A second level of regulation is achieved by varying the frequency with which a gene in the active conformation is transcribed into RNA by an RNA polymerase. There is evidence for regulation of RNA synthesis at both of these levels—for example, in response to hormone induction. At both levels, protein factors are believed to perform the regulation, for example, by binding to special DNA regions flanking the transcribed gene.

Regulation of RNA after synthesis. After synthesis, RNA molecules undergo selective processing, which results in the export of only a subpopulation of RNA molecules to the cytoplasm. Furthermore, the stability in the cytoplasm of a particular type of mRNA can be regulated. There is more and more evidence for regulation of gene expression by prolonging or curtailing the life of a specific mRNA. For example, the hormone prolactin increases synthesis of milk proteins in tissue by causing a twofold rise in the rate of mRNA synthesis; but it also causes a 17-fold rise in mRNA lifetime, so that, in this case, the main cause of increased protein synthesis is the prolonged availability of mRNA. Conversely, there is evidence for selective destabilization of some mRNA—such as histone mRNA, which is rapidly broken down when DNA replication is interrupted. Finally, there are examples of selective regulation of the translation of mRNA into protein. This appears to be less generally used as a regulatory step than the other levels illustrated, but it is clear that it too can be used to regulate the expression of some genes. (R.A.La.)

The mitochondrion and the chloroplast

Mitochondria and chloroplasts are the powerhouses of the cell. Mitochondria appear in both plant and animal cells as elongated, cylindrical bodies, roughly one micrometre in length and closely packed in regions actively using metabolic energy. Oxidizing the products of cytoplasmic metabolism, they convert the energy so liberated into adenosine triphosphate (ATP), the energy currency of the cell. Chloroplasts are the photosynthetic organelles in plant and plantlike bacterial cells. They trap light energy and convert it partly into ATP but mainly into certain chemically reduced molecules that, together with ATP, are used in the first steps of carbohydrate production. Mitochondria and chloroplasts share a certain structural resemblance, and both have a somewhat independent existence within the cell, synthesizing some proteins and dividing according to their own genetic instructions.

MITOCHONDRIAL AND CHLOROPLASTIC STRUCTURE

Both organelles are bounded by an external porous membrane, which allows passage of small metabolites but which is an effective barrier, in general, to the movement of large cytoplasmic proteins. An inner membrane, concentric to the outer, is impermeable even to small ions, such as protons. The membranes of both organelles have a lipid bilayer construction, which has been shown to be a selective barrier against the passage of charged, water-soluble molecules and ions (see above *The plasma membrane: Chemical composition and structure of the membrane*). Between the inner and outer layers is the intermembrane space.

In mitochondria the inner membrane is thrown into a vast number of folds called cristae, enormously expanding the surface area of the membrane. In contrast, the inner membrane of chloroplasts is relatively smooth, but within this membrane is yet another series of membranes, also enormously folded, that form a set of flattened, disklike sacs called the thylakoids. The organellar space enclosed by the inner membrane is called the matrix in mitochondria and the stroma in chloroplasts. Both spaces are filled with a fluid containing a rich mixture of metabolic products, enzymes, and ions. Enclosed by the thylakoid membrane of the chloroplast is the thylakoid space. The extraordinary chemical capabilities of the two organelles lie in the cristae and the thylakoids. Both membranes are studded with enzymatic proteins either traversing the bilayer or dissolved in the lipid. These proteins contribute to the production of energy by transporting material across the membranes and by serving as electron carriers in the important oxidation–reduction reaction.

METABOLIC FUNCTIONS

The oxidation–reduction reaction

Crucial to the function of mitochondria and chloroplasts is the chemistry of the oxidation–reduction, or redox, reaction. This controlled burning of material comprises the transfer of electrons from one compound, called the donor, to another, called the acceptor. All compounds taking part in redox reactions are ranked in a descending scale according to their ability to act as electron donors. Those higher in the scale donate electrons to their fellows lower down, which have a lesser tendency to donate, but a correspondingly greater tendency to accept, electrons. Each acceptor in turn donates electrons to the next compound down the scale, forming a donor–acceptor chain extending from the greatest donating ability to the least.

At the top of the scale is hydrogen, the most abundant element in the universe. The nucleus of a hydrogen atom is composed of one positively charged proton; around the nucleus revolves one negatively charged electron. In the atmosphere, two hydrogen atoms join to form a hydrogen molecule (H_2). In solution, the two atoms pull apart, and in some chemical reactions the atoms are split into their constituent protons and electrons. In the redox reaction the electrons are passed from one reactant to another. The donation of electrons is called oxidation, and the acceptance is called reduction—hence the descriptive term oxidation–reduction, indicating that one action never takes place without the other.

A hydrogen atom has a great tendency to transfer an electron to an acceptor. An oxygen atom, in contrast, has a great tendency to accept an electron. The burning of hydrogen by oxygen is, chemically, the transfer of an electron from each of two hydrogen atoms to oxygen, so that hydrogen is oxidized and oxygen reduced. The reaction is extremely exergonic; *i.e.,* it liberates much free energy as heat. This is the reaction that takes place within mitochondria, but is so controlled that the heat is liberated not at once but in a series of steps. The free energy, harnessed by the organelle, is coupled to the synthesis of ATP from adenosine diphosphate (ADP) and inorganic phosphate (P_i).

An analogy can be drawn between this controlled reaction and the flow of river water down a lock system. Without the locks, water flow would be rapid and uncontrolled, and no ship could safely ply the river. The locks force water to flow in small, controlled steps conducive to safe navigation. But there is more to a lock system than this. The flow of water down the locks can also be harnessed to raise a ship from a lower to a higher level, with the water rather than the ship expending the energy. In mitochondria, the burning of hydrogen is broken into a series of small, indirect steps following the flow of electrons along a chain of donor-acceptors. Energy is funneled into the chemical bonding of ADP and P_i, raising the free energy of these two compounds to the high level of ATP.

The mitochondrion. *Formation of the electron donors NADH and FADH$_2$.* Through a series of metabolic reactions carried out in the matrix, the mitochondrion converts products of the cell's initial metabolism of fats, amino acids, and sugars into the compound acetyl coenzyme A. The acetate portion of this compound is then oxidized in a chain reaction called the tricarboxylic acid cycle. At the end of this cycle the carbon atoms yield carbon dioxide and the hydrogen atoms are transferred to the cell's

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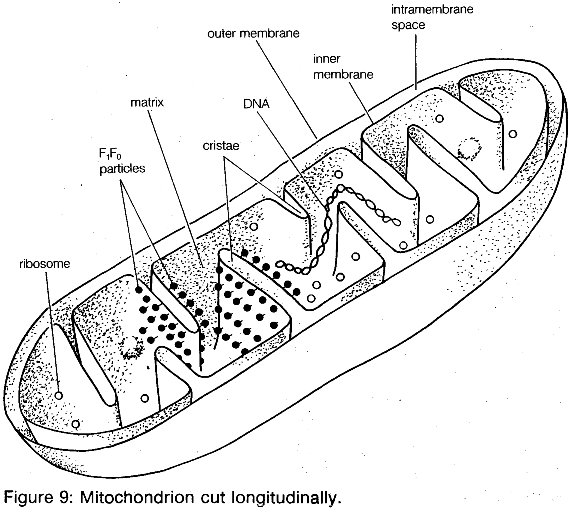

Figure 9: Mitochondrion cut longitudinally.

most important hydrogen acceptors, the coenzymes nicotinamide adenine dinucleotide (NAD$^+$) and flavin adenine dinucleotide (FAD), yielding NADH and FADH$_2$. It is the subsequent oxidation of these hydrogen acceptors that leads eventually to the production of ATP.

NADH and FADH$_2$ are compounds of high electron-donating capacity. Were they to transfer their electrons directly to oxygen, the resulting combustion would release a lethal burst of heat energy. Instead, the energy is released in a series of electron donor-acceptor reactions carried out, within the cristae of the mitochondrion, by a number of proteins and coenzymes that make up the electron-transport, or respiratory, chain.

Membrane proteins as electron carriers

The electron-transport chain. The proteins of this chain are embedded in the cristae membrane, actually traversing the lipid bilayer and protruding from the inner and outer surfaces. The coenzymes are dissolved in the lipid and diffuse through the membrane or across its surface. The proteins are arranged in three large complexes, each composed of a number of polypeptide chains. Each complex is, to continue the hydraulic analogy, a lock in the waterfall of the electron flow and the site at which energy from the overall redox reaction is tapped. The first complex, NADH dehydrogenase, accepts a pair of electrons from the primary electron donor NADH and is reduced in the process. It in turn donates these electrons to the coenzyme ubiquinone, a lipid-soluble molecule composed of a substituted benzene ring attached to a hydrocarbon tail. Ubiquinone, diffusing through the lipid of the cristae membrane, reaches the second large complex of the electron-transport chain, the b-c$_2$ complex, which accepts the electrons, oxidizing ubiquinone and being itself reduced. (This complex can also accept electrons from the second primary electron donor, FADH$_2$, a molecule below NADH in the electron-donating scale.) The b-c$_2$ complex transfers the pair of electrons to cytochrome c, a small protein situated on the outer surface of the cristae membrane. From cytochrome c, electrons pass (four at a time) to the third large complex, cytochrome oxidase, which, in the final step of the chain, transfers the four electrons to two oxygen atoms and two protons, generating two water molecules.

This transfer of electrons, from member to member of the electron-transport chain, provides energy for the synthesis of ATP through an indirect route. At the beginning of the electron-transport chain, NADH and FADH$_2$ split hydrogen atoms into protons and electrons, transferring the electrons to the next protein complex and releasing the protons into the mitochondrial matrix. When each protein complex in turn transfers the electrons down the chain, it uses the energy released in this process to pump protons across the inner membrane into the intermembrane space. (For the dynamics of this pumping action, see above *The plasma membrane: Transport across the membrane.*) This transport of positively charged protons into the intermembrane space, opposite the negatively charged electrons in the matrix, creates an electrical potential that tends to draw the protons back across the membrane. A high concentration of protons outside the membrane also creates the conditions for their diffusion back into the matrix. However, as is explained above, the inner membrane is extremely impermeable to protons. In order for the protons to flow back down the electrochemical gradient, they must traverse the membrane through transport molecules similar to the protein complexes of the electron-transfer chain. These molecules are the so-called F$_1$F$_0$ATPase, a complex protein that, transporting protons back into the matrix, uses the energy released to synthesize ATP. The protons then join the electrons and oxygen atoms to form water. (For further discussion of ATP production, see above *The nature and function of cells: The cell as a collection of self-replicating catalysts: Coupled chemical reactions.*)

This complex chain of events, the basis of the cell's ability to derive ATP from metabolic oxidation, was conceived in its entirety by the British biochemist Peter Mitchell in 1961. The years following the announcement of his chemiosmotic theory saw its ample substantiation and revealed its profound implications for cell biology.

The chemiosmotic theory. The chemiosmotic theory has four postulates:

(1) The inner mitochondrial membrane is impermeable to protons, hydroxide ions, and other cations and anions. This postulate was validated when it was shown that substances allowing protons to flow readily across mitochondrial membranes uncouple oxidative electron transport from ATP production.

(2) Transfer of electrons down the electron-transport chain brings about pumping of protons across the inner membrane, from matrix to intermembrane space. This was demonstrated in the laboratory by a reconstitution of components of the electron-transport chain in artificial membrane vesicles. The stimulation of electron transport caused a measurable buildup of protons within the vesicle.

ATP-synthesizing proteins

(3) The flow of protons down a built-up electrochemical gradient occurs through a proton-dependent ATPase, so that ATP is synthesized from ADP and P$_i$ whenever protons move through the enzyme. This hypothesis was confirmed by the discovery of what came to be known as the F$_1$F$_0$ATPase. Shaped like a knob attached to the membrane by a narrow stalk, F$_1$F$_0$ATPase covers the inner surface of the cristae. Its stalk (the F$_0$ portion) penetrates the lipid bilayer of the inner membrane and is capable of catalyzing the transport of protons. The knob (the F$_1$ portion) is capable of synthesizing as well as splitting, or hydrolyzing, ATP. F$_1$F$_0$ATPase is therefore reversible, either using the energy of proton diffusion to combine ADP and P$_i$ or using the energy of ATP hydrolysis to pump protons out of the matrix.

(4) The inner membrane of the mitochondrion possesses a complement of proteins that bring about the transport of essential metabolites. Numerous carrier systems have been demonstrated to transport into the mitochondrion the products of metabolism that are transformed into substrates for the electron-transport chain. Best known is the ATP–ADP exchange carrier of the inner membrane. Neither ATP nor ADP, being large, charged molecules, can cross the membrane unaided, but ADP must enter and ATP must leave the mitochondrial matrix in order for ATP synthesis to continue. A single protein conducts the counter-transport of ATP against ADP, the energy released by the flow of ATP down its concentration gradient being coupled to the pumping of ADP up its gradient and into the mitochondrion.

The chloroplast. *Trapping of light.* Light travels as packets of energy known as photons and is absorbed in this form by light-absorbing chlorophyll molecules embedded in the thylakoid membrane of the chloroplast. The chlorophyll molecules are grouped into clusters of several hundred, called antenna complexes, by specific proteins that hold them onto the thylakoid membrane and confer

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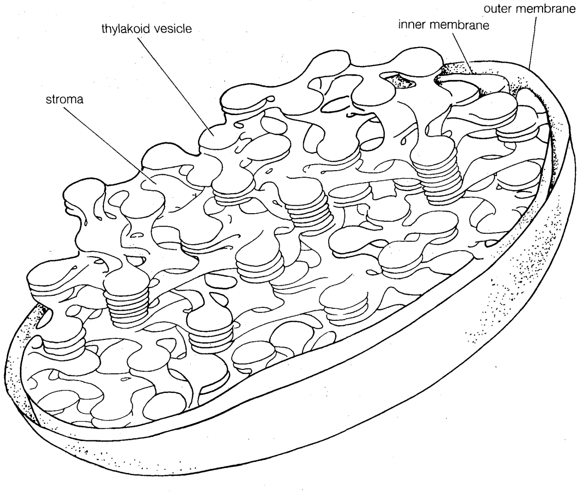

Figure 10: *The structure of a chloroplast.*
The internal (thylakoid) membrane vesicles are organized into stacks, which reside in a matrix called the stroma. All the chlorophyll in the chloroplast is contained in the membranes of the thylakoid vesicles.

upon one of them the special property of being a reaction centre. Into this centre are funneled photons absorbed by the other chlorophylls of the antenna. The energy of the photon is absorbed by an electron of the reaction centre molecule in sufficient quantity to enable its acceptance by a nearby coenzyme, which cannot accept electrons at low energy levels. This coenzyme has a high electron-donor capability; it initiates the transfer of the electron down an electron-transport chain similar to that of the mitochondrion. Meanwhile, the loss of the negatively charged electron leaves a positively charged "hole" in the reaction centre chlorophyll molecule. This hole is filled by the enzymatic splitting of water into molecular oxygen, protons, and electrons and the transfer of an electron to the chlorophyll. The oxygen is released by the chloroplast, making its way out of the plant and into the atmosphere. The protons, in a process similar to that in the mitochondrion, are pumped through the thylakoid membrane and into the thylakoid space. Their facilitated diffusion back into the stroma through proteins embedded in the membrane powers the synthesis of ATP.

Energizing of electrons

This part of the photosynthetic process is called photosystem II. At the end of the electron-transport chain of this system is another reaction centre molecule. The electron is again energized by photons and then transported down another chain, which makes up photosystem I. This system uses the energy released in electron transfer to join a proton to nicotinamide adenine dinucleotide phosphate ($NADP^+$), a phosphorylated derivative of NAD^+, forming NADPH. NADPH is a high-energy electron donor that, with ATP, fuels the conversion of carbon dioxide into the carbohydrate foods of the plant cell.

Fixation of carbon dioxide. NADPH remains within the stroma of the chloroplast for use in the fixation of carbon dioxide (CO_2). In a complex cycle of chemical reactions, CO_2 is bound to a five-carbon ribulose biphosphate compound. The six-carbon intermediate so formed is then split into three-carbon phosphoglycerate. With energy supplied by the breakdown of NADPH and ATP, this compound is eventually formed into glyceraldehyde 3-phosphate, an important sugar intermediate of metabolism. One glyceraldehyde molecule is exported from the chloroplast, for further conversion in the cytoplasm, for every five that undergo an ATP-powered re-formation into the five-carbon ribulose biphosphate. In this way three molecules of CO_2 yield one molecule of glyceraldehyde 3-phosphate, while the entire fixation cycle hydrolyzes nine molecules of ATP and oxidizes six molecules of NADPH.

EVOLUTIONARY ORIGINS

The mitochondrion and chloroplast as independent entities. In addition to their remarkable metabolic capabilities, both mitochondria and chloroplasts synthesize on their own a number of proteins and lipids necessary for their structure and activity. Not only do they contain the machinery necessary for this, but they also possess the genetic material to direct it. DNA within these organelles has a circular structure reminiscent of bacterial, not eukaryotic, DNA. Also as in bacteria, the DNA is not associated with histones. Along with the DNA are protein-synthesizing ribosomes, of bacterial rather than eukaryotic size, that are sensitive to bacteria-inhibiting antibiotics.

Similarities to bacteria

Only a small portion of the mitochondrion's total number of proteins is synthesized within the organelle. At least 90 proteins are encoded and made in the human cytoplasm specifically for export to the mitochondrion, while the mitochondrial DNA encode only 13 different proteins. The proteins that do contain components synthesized within the mitochondrion often possess, in addition, components synthesized in the cytoplasm. Mitochondrial and chloroplastic proteins synthesized in the cytoplasm have to enter the organelle by a complex process, crossing one or more membranes. Why this is possible for most, but not all, proteins is not known.

The endosymbiont hypothesis. Mitochondria and chloroplasts are self-dividing; they contain their own DNA and protein-synthesizing machinery, all of the bacterial type. Chloroplasts produce ATP and trap photons by mechanisms that are complex and yet similar to those of certain bacteria. These phenomena have led to the hypothesis that the two organelles are direct descendants of bacteria that entered primitive nucleated cells in a number of infections. Among billions of such infective events, a few could have led to the development of stable, symbiotic associations between nucleated hosts and bacterial parasites. The hosts would provide the parasites with a stable osmotic environment and easy access to nutrients, and the bacteria would repay by providing an effective oxidative ATP-producing system or a photosynthetic energy-producing reaction. The further development of genetic techniques may make it possible to identify the classes of bacteria that took part in these putative events. (W.D.S.)

The cytoskeleton

The cytoskeleton is the name given to a fibrous network formed by different types of long protein filaments present throughout the cytoplasm of eukaryotic cells (cells containing a nucleus). As the term cytoskeleton implies, these filaments create a scaffold or framework that organizes other cell constituents and maintains the shape of the cell. In addition, some filaments cause coherent movements, both of the cell itself and of its internal organelles. Prokaryotic (non-nucleated) cells, which are much smaller than eukaryotic cells, seem not to require an internal skeleton of this type.

Three major types of cytoskeletal filaments are commonly recognized: actin filaments, microtubules, and intermediate filaments. Actin filaments and microtubules are dynamic structures that continuously assemble and disassemble in most cells. Intermediate filaments are stabler and seem to be involved mainly in reinforcing cell structures. A wide variety of accessory proteins work in concert with each type of filament, linking filaments to one another and to the plasma membrane and helping to form the networks that endow the cytoskeleton with its unique functions. Many of these accessory proteins have not yet been characterized, limiting understanding of the cytoskeleton.

ACTIN FILAMENTS

Actin is a small globular protein molecule that polymerizes to form long filaments. Because each actin subunit faces in the same direction, the actin filament is polar, with different "plus" and "minus" ends. An abundant protein in all eukaryotic cells, actin is most easily studied in muscle cells, where it assembles into unusually stable filaments. In muscle the actin filaments are organized into regular arrays, and they interdigitate with a set of thicker filaments formed from a second protein called myosin. These two proteins create the force responsible for muscle contraction. When the signal to contract is sent along a nerve to the muscle, the myosin molecules are activated to hydrolyze adenosine triphosphate (ATP), releasing energy in such a way that a myosin filament moves along an actin filament, causing the two filaments to slide past each other. Large muscles are composed of bundles of many long muscle cells; when the actin–myosin assemblies slide, each of these giant muscle cells shortens, and the overall effect is the contraction of the entire muscle.

Muscle contraction

Actin is also present in non-muscle cells, where it forms less highly ordered arrays of filaments responsible for certain types of cellular movement. Actin is intimately involved in linking the plasma membrane to the underlying cytoplasm. In some cells, these filaments are stabilized by the binding of accessory proteins so as to form microvilli, stable protrusions of the plasma membrane that resemble tiny bristles. Microvilli on the surface of epithelial cells function to increase the cell's surface area, facilitating the absorption of vital molecules through the membrane. Other types of microvilli are involved in the detection of sound in the ear, where their movement, caused by sound waves, sends an electrical signal to the brain.

Many actin filaments in non-muscle cells have only a transient existence, polymerizing and depolymerizing in controlled ways that create movement. For example, many cells continually send out and retract tiny "microspikes," long, needlelike projections of the plasma membrane that are thought to enable cells to "taste" their environment. Like microvilli, microspikes are formed when actin filaments push out the membrane, but because these actin filaments are less organized and less stable, microspikes have only a brief existence. Another actin structure only transiently associated with the plasma membrane is the contractile ring, which is composed of actin filaments running around the circumference of the cell during cell division. As its name implies, this ring pulls in the plasma membrane by a myosin-dependent process, thereby pinching the cell in half. How actin is specifically organized into these different transient arrays is unknown, but actin-binding proteins on the plasma membrane probably play a central role.

MICROTUBULES

Movement of cilia and flagella

Microtubules are long filaments formed from 13 strands of a globular subunit called tubulin, with the strands arranged in the form of a hollow cylinder. Like actin filaments, microtubules are polar, having "plus" and "minus" ends. Most microtubules are constantly assembling and disassembling, but stable microtubules are found in cilia and flagella. Cilia are hairlike motile structures found on the surface of certain types of epithelial cells, where they act as little propellers to move fluid and particles over the cell surface. They are closely related in structure to flagella such as the sperm tail, the whiplike motion of which enables a sperm cell to propel itself rapidly through fluids. In all of these motile structures, a bundle of microtubules is connected in a regular array by various accessory proteins that act as links and spokes in the assembly. Movement of the cilia or flagella occurs when adjacent microtubules slide past one another, bending the structures. This motion is caused by the protein dynein, which uses the energy of ATP hydrolysis to move along the microtubules, in a manner resembling the movement of myosin along actin filaments.

In most cells, microtubules grow outward, from the cell centre to the plasma membrane, from a special region of the cytoplasm near the nuclear envelope called the centrosome. Experiments show that the "minus" ends of these microtubules are embedded in the centrosome and that the "plus" ends terminate near the plasma membrane. Just before cell division, the centrosome divides in two. The two halves move to opposite sides of the nuclear envelope, each carrying a starlike array of microtubules, and become the two poles of the mitotic spindle, which has a central role in partitioning the chromosomes and other cell components into two daughter cells during mitosis (see below *Cell division and growth*).

INTERMEDIATE FILAMENTS

Intermediate filaments have a larger diameter than actin filaments and a smaller diameter than microtubules. Unlike the other two types of filament, intermediate filaments are formed from elongated, ropelike subunits of a variety of different types. Their best-understood roles are as keratin filaments, strengthening cell sheets such as the skin, and as the nuclear lamina, a fibrous shell that underlies and supports the nuclear membrane.

STRUCTURAL RELATION OF THE FILAMENTS

Microtubules and actin filaments are thought to form the basic foundation of the cytoskeleton. The microtubules are believed to be the overall organizers of the cytoskeleton, affecting the position and function of both actin filaments and intermediate filaments. In addition, the movements of certain organelles depend on microtubules; specific proteins that bind to microtubules and hydrolyze ATP seem to be responsible for moving these organelles.

Difficulty of studying the cytoskeleton

An important unresolved issue in cell biology concerns the extent to which the many different soluble proteins in the cytoplasm are spatially organized into different regions. Such organization is presumably mediated in large part by the binding of metabolic enzymes and other proteins to a network of proteins linked in turn to different parts of the cytoskeleton. The dynamic nature of the cytoskeleton makes the nature of this organization difficult to clarify. (B.M.A.)

The cell matrix and cell-to-cell communication

The development of single cells into multicellular organisms involves a number of adaptations. The cells become specialized, acquiring distinct functions that contribute to the survival of the organism. The behaviour of individual cells is also integrated with that of similar cells, so that they act together in a regulated fashion. To achieve this integration, cells assemble into specialized tissues, each tissue being composed of cells and the spaces outside of the cells.

The surface of cells is important in coordinating their activities within tissues. Embedded in the plasma membrane of each cell are a number of proteins that interact with the surface or secretions of other cells. These proteins enable cells to "recognize" and adhere to the extracellular matrix and one another and to form populations distinct from surrounding cells. These interactions are keys to the organizational behaviour of cell populations. Each contributes to the formation of embryonic tissues, and each is necessary for normal tissue function in the adult organism.

THE EXTRACELLULAR MATRIX

A substantial part of tissues is the space outside of the cells, called the extracellular space. This is filled with a composite material, known as the extracellular matrix, composed of a gel in which are suspended a number of fibrous proteins. The gel consists of large polysaccharide (complex sugar) molecules in a water solution of inorganic salts, nutrients, and waste products known as the interstitial fluid. The major types of protein in the matrix are structural proteins and adhesive proteins.

Effects of matrix components on tissue properties

There are two general types of tissue distinct not only in their cellular organization but also in the composition of their extracellular matrix. The first type, mesenchymal tissue, is made up of clusters of cells grouped together but not closely adherent to one another. They synthesize a highly hydrated gel, rich in salts, fluid, and fibres, known as the interstitial matrix. Connective tissue is a mesenchyme that fastens together other more highly organized tissues. The solidity of various connective tissues varies according to the consistency of their extracellular matrix, which in turn depends on the water content of the gels, the amount and type of polysaccharides and structural proteins, and the presence of other salts. For example, bone is rich in calcium phosphate, giving this tissue its rigidity; tendons are mostly fibrous structural proteins, yielding a ropelike consistency; and joint spaces are filled with a lubricating fluid of mostly polysaccharide and interstitial fluid.

Epithelial tissues, the second type, are sheets of cells adhering at their side, or lateral, surfaces. They synthesize and deposit at their bottom, or basal, surfaces an organized complex of matrix materials known as the basal lamina or basement membrane. This thin layer serves as a boundary with connective tissue and as a substrate to which epithelial cells are attached (see Figure 11).

Matrix polysaccharides. The polysaccharides, or glycans, of the extracellular matrix are responsible for its gellike quality and for organizing its components. These large, acidic molecules exist alone (as glycosaminoglycans) or in combination with small proteins (as proteoglycans). They bind an extraordinarily large amount of water, thus forming massively swollen gels that fill the spaces between cells. Bound to proteins, they also organize other molecules in the extracellular matrix. The firmness and resiliency of cartilage, as at the surface of joints, is due to highly organized proteoglycans that bind water tightly.

Matrix proteins. Matrix proteins are large molecules tightly bound to form extensive networks of insoluble fibres. These fibres may even exceed the size of the cells themselves. The proteins are of two general types, structural and adhesive.

Properties of collagen and elastin

The structural proteins, collagen and elastin, are the dominant matrix proteins. At least 10 different types of collagen are present in various tissues. The most common, type I collagen, is the most abundant protein in vertebrate animals, accounting for nearly 25 percent of the total protein in the body. The various collagen types share structural features, all being composed of three in-

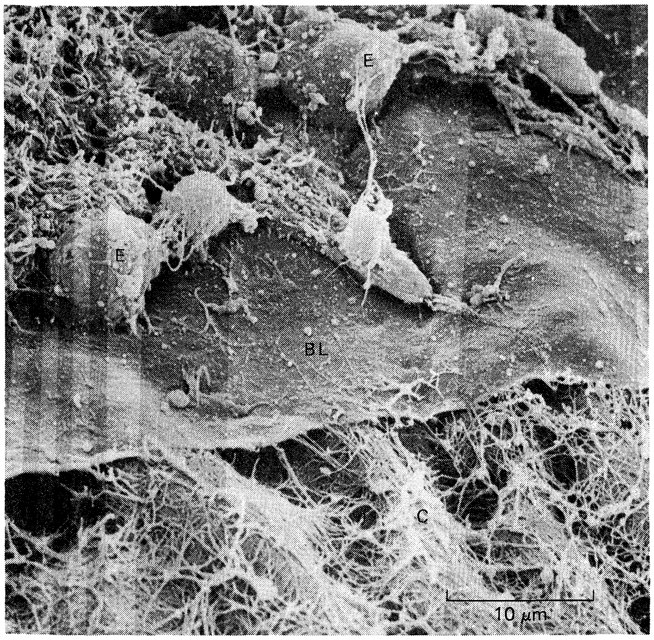

Figure 11: *Scanning electron micrograph of basal lamina in the cornea of a chick embryo.*
Some of the epithelial cells (E) have been removed to expose the upper surface of the basal lamina (BL). Note the network of collagen fibrils (C) interacting with the lower face of the lamina.
By courtesy of Robert Trelstad

tertwined polypeptide chains. In some collagens the chains are linked together by covalent bonds, yielding a ropelike structure of great tensile strength. Indeed, the toughness of leather, chemically treated animal skin, is due to its content of collagen. Elastin is also a cross-linked protein, but instead of forming rigid coils, it imparts elasticity to tissues. Only one type of elastin is known; it varies in elasticity according to variations in its cross-linking.

The adhesive proteins of the extracellular matrix bind matrix molecules to one another and to cell surfaces. These proteins are modular, in that they contain several functional domains packaged together in a single molecule. Each domain binds to a specific matrix component or to a specific site on a cell. The major adhesive protein of the interstitial matrix is called fibronectin; that of the basal lamina is known as laminin.

Cell–matrix interactions. Molecules intimately associated with the plasma membrane link cells to the extracellular matrix. These molecules, called matrix receptors, bind selectively to specific matrix components and interact, directly or indirectly, with actin protein fibres that form the cytoskeleton inside the cell. This association of actin fibres with matrix components via receptors on the cell membrane can influence the organization of membrane molecules as well as matrix components and can modify the shape and function of the cytoskeleton. Changes in the cytoskeleton can lead to changes in cell shape, movement, metabolism, and development.

INTERCELLULAR RECOGNITION AND CELL ADHESION

The simplest examples of the ability of cells to recognize and adhere to one another come from studies of organisms that live much of their lives as single cells but form a multicellular aggregate for a specific purpose. For example, when starved, several types of single-cell organisms band together to develop the specialized cells needed for reproduction. In this process, certain cells at the centre of the developing aggregate secrete chemicals that cause the other cells to adhere tightly into a group. In the case of slime mold amoebas, starvation causes the secretion of a compound, cyclic adenosine monophosphate (cyclic AMP or CAMP), that induces the cells to stick together end-to-end. With further aggregation, the cells produce another cell-surface glycoprotein with which they stick to one another over their entire surfaces. The cellular aggregates then produce an extracellular matrix, which holds the cells together in a specific structural form.

Tissue and species recognition. Some multicellular animals or tissues can be dissociated into suspensions of single cells that show the same cellular recognition and adhesion as do aggregates of single-cell organisms. The marine sponge, for example, can be sieved through a mesh, yielding single cells and cells in clumps. When this cell suspension is rotated in culture, the cells reaggregate and in time reform a normal sponge. This reassociation shows selective cell recognition; that is, only cells of the same species reassociate. The ability of the cells to distinguish cells of their own species from those of others is mediated by proteoglycan molecules in the extracellular matrix. The proteoglycan binds to specific cell-surface receptor sites that are unique to a single species of sponge.

Spontaneous reassociation of dissolved tissue

Cells from tissues of vertebrate animals can, like sponge cells, be dissociated and allowed to reaggregate. These cells, however, show tissue specificity rather than species specificity. For example, when vertebrate embryonic cells from two different tissues are dissociated and then rotated together in culture, the cells form a multicellular aggregate within which they sort according to the type of tissue. This sorting occurs regardless of whether the cells are from the same or different species, indicating that tissue specificity is more important in adhesive recognition than species specificity. The specificity is due to a set of cell-surface glycoproteins called cell adhesion molecules (CAM). A portion of the CAM that extends from the surface of a cell adheres to identical molecules on the surface of adjacent cells. These CAM appear early in embryonic life, and their amounts in tissues change as the organs develop. The CAM, however, are not responsible for the stable adhesion of one cell to another; this more permanent adhesion is carried out by cell junctions.

Cell junctions. There are three functional categories of cell junction: adhering junctions, often called desmosomes; tight, or occluding, junctions; and gap, or permeable, junctions. Adhering junctions hold cells together mechanically and are associated with intracellular fibres of the cytoskeleton. Tight junctions also hold cells together, but they form a nearly leak-proof intercellular seal by fusion of adjacent plasma membranes. Both adhering junctions and tight junctions are present primarily in epithelial cells. On the other hand, all cell types possess gap junctions, which allow small molecules to pass from one cell to the next through a channel.

Adhering junctions. Cells subject to abrasion or other mechanical stress, such as those of the surface epithelia of the skin, have junctions that adhere cells to one another and to the extracellular matrix. These adhering junctions are called desmosomes when occurring between cells and hemidesmosomes (half-desmosomes) when linked to the matrix. Adhering junctions distribute mechanical shear force throughout the tissue and to the underlying matrix by virtue of their association with intermediate filaments crossing the interior of the cell. The linkage of these filaments, also called keratin filaments, to the desmosomes and, through these junctions, to adjacent cells provides a nearly continuous fibrous network throughout an epithelial sheet. Adhering junctions are also seen in other types of cells—for example, in the muscles of the heart and uterus—allowing these cells to remain anchored together despite the contractions of the muscles.

Tight junctions. Sheets of cells separate fluids within the organ from fluids outside, as in the epithelial layer lining the intestine. This separation requires leak-proof junctions between cells. Tight junctions form leak-proof seals by fusing the plasma membranes of adjacent cells, creating a continuous barrier through which molecules cannot pass. The membranes are fused by the melding of their lipid layers, which repel large, water-soluble molecules. In invertebrates, this function is provided by septate junctions, in which the proteins of the membrane rather than the lipids form the seal.

Gap junctions. These junctions allow communication between adjacent cells via the passage of small molecules directly from the cytoplasm of one cell to that of another. Molecules that can pass between cells coupled by gap junctions include inorganic salts, sugars, amino acids, nucleotides, and vitamins but not large molecules such as proteins or nucleic acids.

Passage of electrical current

Gap junctions are crucial to the integration of certain cellular activities. For example, muscle cells generate electrical current by the movement of inorganic salts. If the cells are coupled, they will share this electrical current, allowing the synchronous contraction of all the cells in the tissue. This coupling function requires the regulation of molecular traffic through the gaps. The junctions are not open pores but dynamic channels, which change their permeability with changes in cellular activity. They consist of proteins completely crossing the plasma membrane as six-sided columns with central pores. Under certain conditions the proteins change shape, causing the pores to become smaller or larger and thus changing the permeability of the junction.

Gap junctions are also found in tissues that are not electrically active. In these tissues, the junctions allow nutrients and waste products to travel throughout the tissue. Cells in such tissues are said to be metabolically coupled. During the formation of embryos, gap junctions are crucial to establishing differences between separate groups of cells, the coupled cells undergoing development together to become a specialized tissue.

CELL-TO-CELL COMMUNICATION VIA CHEMICAL SIGNALING

In addition to cell–matrix and cell–cell interactions, cell behaviour in multicellular organisms is coordinated by the passage of chemical or electrical signals between cells. The most common form of chemical signaling is via molecules secreted from the cells and moving through the extracellular space. Signaling molecules may also remain on cell surfaces, influencing other cells only after the cells make physical contact. Finally, as noted above, gap junctions allow small molecules to move between the cytoplasms of adjacent cells.

Types of chemical signaling. Chemical signals secreted by cells can act over varying distances. In the autocrine signaling process, molecules act on the same cells that produce them. In paracrine signaling, they act on nearby cells. Autocrine signals include extracellular matrix molecules and various factors that stimulate cell growth. An example of paracrine signals is the chemical transmitted from nerve to muscle that causes the muscle to contract. In this instance, the muscle cells have regions specialized to receive chemical signals from an adjacent nerve cell. In both autocrine and paracrine signaling, the chemical signal works in the immediate vicinity of the cell that produces it and is present at high concentrations. A chemical signal picked up by the bloodstream and taken to distant sites is called an endocrine signal. Most hormones produced in vertebrates are endocrine signals, such as the hormones produced in the pituitary gland at the base of the brain and carried by the bloodstream to act at low concentrations on the thyroid or adrenal glands.

Endocrine signaling system

The concentration at which a chemical signal acts has significance for its target cell. Chemical signals that act at high concentration act locally and rapidly. On the other hand, chemical signals that act at low concentrations act at distances and are generally slow.

Signal receptors. The ability of a cell to respond to an extracellular signal depends on the presence of specific proteins, called receptors, on its surface or in the cytoplasm. The receptors bind the chemical signals tightly, ultimately triggering a mechanism that modifies the behaviour of the target cell. Cells may contain an array of specific receptors allowing them to respond to a variety of chemical signals.

Signal molecules are either soluble or insoluble, this feature determining how they interact with the cell. Water-soluble molecules such as the polypeptide hormone insulin bind to receptors at cell surfaces. On the other hand, lipid-soluble molecules, such as the steroid hormones produced by the ovary or testis, pass through the lipid bilayer of the plasma membrane to reach receptors within the cell. Extracellular matrix molecules are chemical signals, but, because of their size and insolubility, they act only on cell surface receptors and are neither taken up by the cells nor rapidly destroyed.

Cellular response. The binding of chemical signals to their corresponding receptors induces events within the cell that ultimately change its behaviour. The nature of these intracellular events differs according to the type of receptor. Also, the same chemical signal causes different responses in different types of cell.

Cell surface receptors work in several ways when they are occupied. Some enter the cell still bound to the chemical signal. Others activate membrane enzymes, which produce certain intracellular chemical mediators. Still other receptors open membrane channels, allowing a flow of ions that causes either a change in the electrical properties of the membrane or a change in the ion concentration in the cytoplasm. This regulation of enzymes or membrane channels produces changes in the concentration of intracellular signaling molecules, which are often called second messengers (the first messenger being the extracellular chemical signal bound to the receptor).

Two common chemical signals

Two common intracellular signaling molecules are cyclic AMP and the calcium ion. Cyclic AMP is a derivative of adenosine triphosphate, the ubiquitous energy-carrying molecule of the cell. The intracellular concentrations of both cyclic AMP and calcium ions are normally very low. The binding of an extracellular chemical signal to a cell surface receptor may stimulate an enzyme complex in the membrane to produce cyclic AMP. This second messenger then diffuses into the cytoplasm and acts on intracellular enzymes that modify the behaviour of the cell—for example, by causing it to increase protein synthesis. The action of cyclic AMP is brief because it is rapidly degraded by specific enzymes.

Occupancy of other surface receptors causes a transient opening of membrane channels. This can allow calcium ions to enter the cytoplasm from the extracellular space, where its concentration is higher. The action of calcium ions is also brief because they are rapidly pumped out of the cell or bound to intracellular molecules, lowering the cytoplasmic concentration to the state existing before stimulation.

Some extracellular chemical signals enter the cell intact, still bound to the receptor, without generating a second messenger. In this mechanism, receptor occupancy causes individual receptors within the plasma membrane to aggregate spontaneously. That portion of the membrane containing the aggregated receptors is then taken into the cell, where it fuses with various membrane-bounded organelles in the cytoplasm. In some instances the chemical signal is released within the organelles, and in almost all instances the ingested membrane is rapidly returned to the plasma membrane along with the surface receptors.

(M.R.B.)

THE PLANT CELL WALL

The plant cell wall is a specialized form of extracellular matrix that surrounds every cell of a plant and is responsible for many of the characteristics distinguishing plant from animal cells. Although often perceived as an inactive product serving mainly mechanical and structural purposes, the cell wall actually has a multitude of functions upon which plant life depends. Such functions include: (1) providing the protoplast, or living cell, with mechanical protection and a chemically buffered environment; (2) providing a porous medium for the circulation and distribution of water, minerals, and other small nutrient molecules; (3) providing rigid building blocks from which stable structures of higher order, such as leaves and stems, can be produced; and (4) providing a storage site of regulatory molecules that sense the presence of pathogenic microbes and control the development of tissues.

Mechanical properties of wall layers. All cell walls contain two layers, the middle lamella and the primary cell wall, and many cells produce an additional layer, called the secondary wall. The middle lamella serves as a cementing layer between the primary walls of adjacent cells. The primary wall is the cellulose-containing layer laid down by cells that are dividing and growing. To allow for cell wall expansion during growth, primary walls are thinner and less rigid than those of cells that have stopped growing. A fully grown plant cell may retain its primary

From primary to secondary wall

cell wall (sometimes thickening it), or it may deposit an additional, rigidifying layer of different composition; this is the secondary wall. Secondary cell walls are responsible for most of the plant's mechanical support as well as the mechanical properties prized in wood. In contrast to the permanent stiffness and load-bearing capacity of thick secondary walls, the thin primary walls are capable of serving a structural, supportive role only when the vacuoles within the cell are filled with water to the point that they exert a turgor pressure against the cell wall. Turgor-induced stiffening of primary walls is analogous to the stiffening of the sides of a pneumatic tire by air pressure. The wilting of flowers and leaves is caused by a loss of turgor pressure, which results in turn from the loss of water from the plant cells.

Components of the cell wall. Although primary and secondary wall layers differ in detailed chemical composition and structural organization, their basic architecture is the same, consisting of cellulose fibres of great tensile strength embedded in a water-saturated matrix of polysaccharides and structural glycoproteins.

Cellulose. Cellulose consists of several thousand glucose molecules linked end to end. The chemical links between the individual glucose subunits give each cellulose molecule a flat, ribbonlike structure that allows adjacent molecules to band laterally together into microfibrils with lengths ranging from two to seven micrometres. Cellulose fibrils are synthesized by enzymes floating in the plasma membrane of the cell and arranged in a rosette configuration. Each rosette appears capable of "spinning" a microfibril into the cell wall. During this process, as new glucose subunits are added to the growing end of the fibril, the rosette is pushed around the cell on the surface of the plasma membrane and its cellulose fibril becomes wrapped around the protoplast. Thus, each plant cell can be viewed as making its own cellulose fibril cocoon.

Matrix polysaccharides. The two major classes of cell wall matrix polysaccharides are the hemicelluloses and the pectic polysaccharides, or pectins. Both are synthesized in the Golgi apparatus, brought to the cell surface in small vesicles, and secreted into the cell wall.

Hemicelluloses consist of glucose molecules arranged end to end as in cellulose, with short side chains of xylose and other uncharged sugars attached to one side of the ribbon. The other side of the ribbon binds tightly to the surface of cellulose fibrils, thereby coating the microfibrils with hemicellulose and preventing them from adhering together in an uncontrolled manner. Hemicellulose molecules have been shown to regulate the rate at which primary cell walls expand during growth.

Action of pectins

The heterogenous, branched, and highly hydrated pectic polysaccharides differ from hemicelluloses in important respects. Most notably, they are negatively charged owing to galacturonic acid residues, which, together with rhamnose sugar molecules, form the linear backbone of all pectic polysaccharides. The backbone contains stretches of pure galacturonic acid residues interrupted by segments in which galacturonic acid and rhamnose residues alternate; attached to these latter segments are complex, branched sugar side chains. Because of their negative charge, pectic polysaccharides bind tightly to positively charged ions, or cations. In cell walls, calcium ions cross-link the stretches of pure galacturonic acid residues tightly, while leaving the rhamnose-containing segments in a more open, porous configuration. This cross-linking creates the semirigid gel properties characteristic of the cell wall matrix—a process exploited in the preparation of jellied preserves.

Proteins. Although plant cell walls contain only small amounts of protein, they serve a number of important functions. The most prominent group are the hydroxyproline-rich glycoproteins, shaped like rods with connector sites, of which extensin is a prominent example. Extensin contains 45 percent hydroxyproline and 14 percent serine residues distributed along its length. Every hydroxyproline residue carries a short side chain of arabinose sugars, and most serine residues carry a galactose sugar; this gives rise to long molecules, resembling bottle brushes, that are secreted into the cell wall toward the end of primary-wall formation and become covalently cross-linked into a mesh at the time that cell growth stops. Plant cells apparently control their ultimate size by regulating the time at which this cross-linking of extensin molecules occurs.

In addition to the structural proteins, cell walls contain a variety of enzymes. Most notable are those that cross-link extensin, lignin, cutin, and suberin molecules into networks. Other enzymes help protect plants against fungal pathogens by breaking fragments off of the cell walls of the fungi. The fragments in turn induce defense responses in underlying cells. The softening and dropping of ripe fruit and of leaves in the autumn are brought about by cell wall-degrading enzymes.

Plastics. Cell wall plastics such as lignin, cutin, and suberin all contain a variety of organic compounds cross-linked into tight, three-dimensional networks that strengthen cell walls and make them more resistant to fungal and bacterial attack. Lignin is the general name for a diverse group of polymers of aromatic alcohols. Deposited mostly in secondary cell walls and providing the rigidity of terrestrial vascular plants, it accounts for up to 30 percent of a plant's dry weight. The diversity of cross-links between the polymers—and the resulting tightness—makes lignin a formidable barrier to the penetration of most microbes.

Cutin and suberin are complex biopolyesters composed of fatty acids and aromatic compounds. Cutin is the major component of the cuticle, the waxy, water-repelling surface layer of cell walls exposed to the environment above ground. By reducing the wetability of leaves and stems—and thereby affecting the ability of fungal spores to germinate—it plays an important part in the defense strategy of plants. Suberin serves with waxes as a surface barrier of underground parts. Its synthesis is also stimulated in cells close to wounds, thereby sealing off the wound surfaces and protecting underlying cells from dehydration.

Intercellular communication. *Plasmodesmata.* Similar to the gap junction of animal cells is the plasmodesma, a channel passing through the cell wall and allowing direct molecular communication between adjacent plant cells. Plasmodesmata are lined with plasma membrane, in effect uniting all connected cells with one continuous cell membrane. Running down the middle of each channel is a thin, membranous tube that connects the endoplasmic reticula (ER) of the two cells. This structure is a remnant of the ER of the original parent cell, which, as the parent cell divided, was caught in the developing cell plate.

Although the precise mechanisms are not fully understood, the plasmodesma is thought to regulate the passage of small molecules such as salts, sugars, and amino acids by constricting or dilating the openings at each end of the channel.

Stimulation of defense responses and cell growth

Oligosaccharides with regulatory functions. The discovery of cell wall fragments with regulatory functions opened a new era in plant research. For years scientists had been puzzled by the chemical complexity of cell wall polysaccharides, which far exceeds the structural requirements of plant cell walls. The answer came when it was found that specific fragments of cell wall polysaccharides, called oligosaccharins, are able to induce specific responses in plant cells and tissues. One such fragment, released by enzymes used by fungi to break down plant cell walls, consists of a linear polymer of 10 to 12 galacturonic acid residues. Exposure of plant cells to such fragments induces them to produce antibiotics known as phytoalexins. In other experiments it has been shown that exposing strips of tobacco stem cells to a different type of cell wall fragment leads to the growth of roots; other fragments lead to the formation of stems, and yet others to the production of flowers. In all instances the concentration of oligosaccharins required to bring about the observed responses is equal to that of hormones in animal cells; indeed, oligosaccharins may be viewed as the oligosaccharide hormones of plants. (L.A.St.)

Cell division and growth

In unicellular organisms, cell division is the means of reproduction; in multicellular organisms, it is the means of tissue growth and maintenance. Survival of the higher organisms depends upon interactions between many cell types, and it is essential that a balanced distribution of types be maintained. This is achieved by the highly regulated process of cell proliferation. The growth and division of different cell populations are regulated in different ways, but the basic mechanisms are similar throughout multicellular organisms.

Division of stem cells and mature cells

Most tissues of the body grow by increasing their cell number, but this growth is highly regulated to maintain a balance between different tissues. In adults most cell division is involved in tissue renewal rather than growth, many types of cells undergoing continuous replacement. Skin cells, for example, are constantly being sloughed off and replaced; in this case the mature differentiated cells do not divide, but their population is renewed by division of immature stem cells. In certain other tissues, such as those of the liver, mature cells remain capable of division to allow growth or regeneration after injury.

In contrast to these patterns, other classes of cell do not divide and are not renewable in the adult organism. These include heart muscle, nerve, and lens cells in mammals. Maintenance and repair of these cells is limited to replacing intracellular components rather than replacing entire cells.

DUPLICATION OF THE GENETIC MATERIAL

Before a cell can divide, it must accurately and completely duplicate the genetic information encoded in its DNA in order for its progeny cells to function and survive. This is a complex problem because of the great length of DNA molecules. Each human chromosome consists of a long double spiral, or helix, each strand of which consists of more than 100,000,000 nucleotides (see above *The nucleus*).

The duplication of DNA is called DNA replication, and it is initiated by complex enzymes called DNA polymerases. These progress along the molecule, reading the sequences of nucleotides that are linked together to make DNA chains. Each strand of the DNA double helix, therefore, acts as a template specifying the nucleotide structure of a new growing chain. After replication, each of the two daughter DNA double helices consists of one parental DNA strand wound around one newly synthesized DNA strand.

In order for DNA to replicate, the two strands must be unwound from each other. Enzymes called helicases unwind the two DNA strands, and additional proteins bind to the separated strands to stabilize them and prevent them from pairing again. In addition, a remarkable class of enzyme called DNA topoisomerase removes the helical twists by cutting either one or both strands and then resealing the cut. These enzymes can also untangle and unknot DNA when it is tightly coiled into a chromatin fibre.

Varying rates of replication

In the circular DNA of bacteria, replication starts at a unique site called the origin of replication, then proceeds in both directions around the molecule until the two processes meet, producing two daughter molecules. In rapidly growing bacteria, a second round of replication can start before the first has finished. The situation in higher organisms is more complicated, as replication moves more slowly than in bacteria. At 500 to 5,000 nucleotides per minute (versus 100,000 nucleotides per minute in bacteria), it would take a human chromosome about a month to replicate if started at a single site. Actually, replication begins at many sites on the long chromosomes of animals, plants, and fungi. Distances between adjacent initiation sites are not always the same; for example, they are closer in the rapidly dividing embryonic cells of frogs or flies than in adult cells of the same species. Less is known about the initiation sites of higher organisms than those of bacteria.

Accurate DNA replication is crucial to ensure that daughter cells have exact copies of the genetic information for synthesizing proteins. Accuracy is achieved by a "proofreading" ability of the DNA polymerase itself. It can erase its own errors and then synthesize anew. There are also repair systems that correct genetic damage to DNA. For example, the incorporation of an incorrect nucleotide, or damage caused by mutagenic agents, can be corrected by cutting out a section of the daughter strand and recopying the parental strand.

CELL DIVISION

Mitosis and cytokinesis. In higher organisms the processes of DNA replication and cell division occur at different times of the cell division cycle. During cell division, DNA condenses to form short, tightly coiled, rodlike chromosomes. Each chromosome then splits longitudinally, forming two identical chromatids. Each pair of chromatids is divided between the two daughter cells during mitosis, or division of the nucleus, a process in which the chromosomes are propelled by attachment to a bundle of microtubules called the mitotic spindle.

From B. Alberts *et al.*, *Molecular Biology of The Cell* (1983); Garland Publishing, Inc.

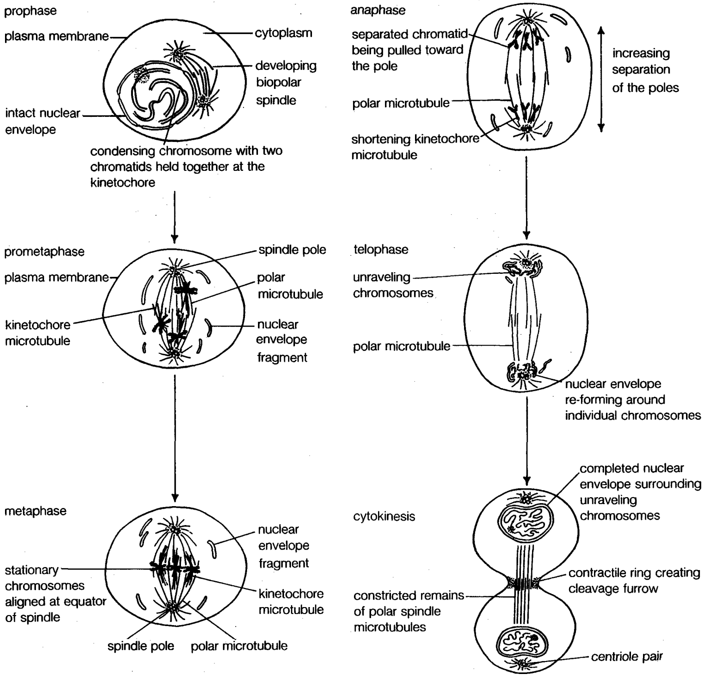

Figure 12: The events during cell division by mitosis.

Mitosis can be divided into five phases (see Figure 12). In prophase the mitotic spindle forms and the chromosomes condense. In prometaphase the nuclear envelope breaks down and the chromosomes attach to the mitotic spindle. Both chromatids of each chromosome attach to the spindle at a specialized chromosomal region called the kinetochore, or centromere. In metaphase the condensed chromosomes align in a plane across the equator of the mitotic spindle. Anaphase follows as the separated chromatids move abruptly toward opposite spindle poles. Finally in telophase a new nuclear envelope forms around each set of unraveling chromatids.

The mitotic spindle

An essential feature of mitosis is the attachment of the chromatids to opposite poles of the mitotic spindle. This ensures that each of the daughter cells will receive a complete set of chromosomes. The mitotic spindle is composed of microtubules, each of which is a tubular assembly of molecules of the protein tubulin (see above *The cytoskeleton*). Some microtubules extend from one spindle pole to the other, while a second class extends from one spindle pole to a chromatid. Microtubules can grow or shrink by the addition or removal of tubulin molecules. This change in length propels attached chromatids to the spindle poles, where they unravel to form new nuclei.

The two poles of the mitotic spindle are occupied by centrosomes, which appear to organize the microtubule arrays. In animal cells each centrosome contains a pair of cylindrical centrioles, which are themselves composed of complex arrays of microtubules. Centrioles duplicate at a precise time in the cell division cycle, usually close to the start of DNA replication. The exact significance of centrioles in cell division is clouded by their absence from plant cells that undergo mitosis. The microtubules in the

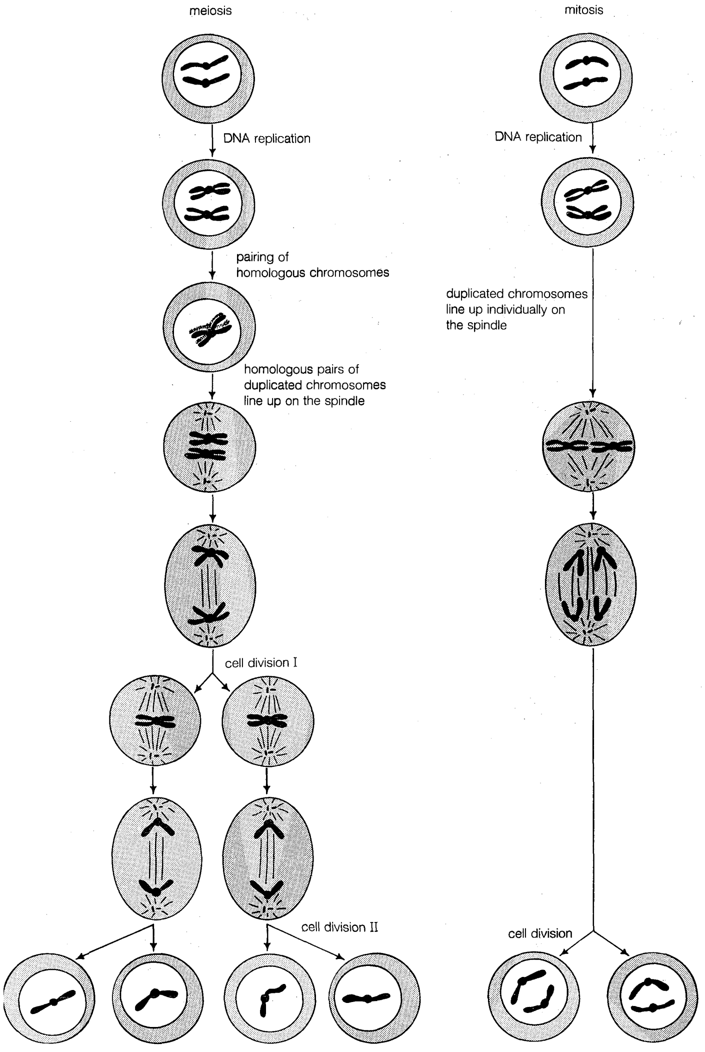

Figure 13: Meiosis compared with a normal mitosis.
From B. Alberts *et al.*, *Molecular Biology of The Cell* (1983); Garland Publishing, Inc.

mitotic spindles of plant cells are less tightly organized, suggesting an organizational role for the centriole.

After mitosis comes cytokinesis, the division of the cytoplasm. This is another process in which animal and plant cells differ. In animal cells cytokinesis is achieved through the constriction of the cell by a ring of contractile microfilaments consisting of actin and probably myosin, the proteins involved in muscle contraction and other forms of cell movement. In plant cells, the cytoplasm is divided by the formation of a new cell wall, called the cell plate, between the two daughter cells. The cell plate arises from small, Golgi-derived vesicles that coalesce in a plane across the equator of the late telophase spindle to form a disk-shaped structure. In this process, each vesicle contributes its membrane to the forming plasma membranes and its matrix contents to the forming cell wall. A second set of vesicles extends the edge of the cell plate until it reaches and fuses with the sides of the parent cell, thereby completely separating the two new daughter cells. At this point, cellulose synthesis commences, and the cell plate becomes a primary cell wall (see above *The cell matrix and cell-to-cell communication: The plant cell wall*).

Meiosis. A specialized division of chromosomes called meiosis occurs during the formation of the reproductive cells, or gametes, of sexually reproducing organisms. Gametes such as ova, sperm, and pollen begin as germ cells, which, like other types of cell, have two copies of each gene in their nuclei. The chromosomes composed of these matching genes are called homologs. During DNA replication, each chromosome duplicates into two attached chromatids. The homologous chromosomes are then separated to opposite poles of the meiotic spindle by tubulin fibres similar to those of the mitotic spindle. At this stage in the meiosis of germ cells, there is a crucial difference from the mitosis of other cells (see Figure 13). In meiosis the two chromatids making up each chromosome remain together, so that whole chromosomes are separated from their homologous partners. Cell division then occurs, followed by a second division that resembles mitosis more closely in that it separates the two chromatids of each remaining chromosome. In this way, when meiosis is complete, each mature gamete receives half of each parental chromosome and, thus, only one copy of each gene instead of the two copies present in other cells. (R.A.La.)

THE CELL DIVISION CYCLE

In bacteria, DNA synthesis can take place uninterrupted between cell divisions, and new cycles of DNA synthesis can begin before previous cycles have finished. In contrast, higher organisms duplicate their DNA exactly once during a discrete period between cell divisions. This period is called the S (for synthetic) phase. It is preceded by a period called G_1 (meaning "first gap") and followed by a period called G_2, during which nuclear DNA synthesis does not occur.

The four periods G_1, S, G_2, and M (for mitosis) make up the cell division cycle. The cell cycle characteristically lasts between 10 and 20 hours in rapidly proliferating adult cells, but it can be arrested for weeks or months in quiescent cells. Prolonged arrest of this type usually occurs during the G_1 phase and is sometimes referred to as G_0. In contrast, some embryonic cells, such as those of vinegar flies (also called fruit flies), can complete entire cycles and divide in only 11 minutes. In these exceptional cases, G_1 and G_2 are undetectable, and mitosis alternates with DNA synthesis. In addition, the duration of the S phase varies dramatically. The vinegar fly embryo takes only four minutes to replicate its DNA, compared with several hours in adult cells of the same species.

Controlled proliferation. Several studies have identified the transition from the G_1 to the S phase as a crucial control point of the cell cycle. Stimuli are known to cause resting cells to proliferate by inducing them to leave G_1 and begin DNA synthesis. These stimuli, called growth factors, are naturally occurring proteins specific to certain groups of cells in the body. They include nerve growth factor, epidermal growth factor, and platelet-derived growth factor. Such factors may have important roles in the healing of wounds as well as in the maintenance and growth of normal tissues. Many growth factors are known to act on the external membrane of the cell, usually by interacting with specialized protein receptor molecules. These respond by triggering further cellular changes, including an increase in calcium levels that makes the cell interior more alkaline and the addition of phosphate groups to the amino acid tyrosine in proteins. Although the complex response of cells to growth factors is not clearly understood, it is known to be of fundamental importance to the control of cell proliferation.

Failure of proliferation control. Cancer can arise when the controlling factors over cell growth fail and allow a cell and its descendants to keep dividing at the expense of the organism. Information on the development of malignant tumours has been obtained by studying viruses that transform cultured cells, leading to malignant behaviour. Transformed cells may differ from their normal progenitors by continuing to proliferate at very high densities, in the absence of growth factors, or in the absence of a solid substrate for support.

Major advances in the understanding of malignant transformation have come from studies of the viral genes that cause transformation. These viral oncogenes have led to the identification of related cellular oncogenes, whose activation can transform cells. Approximately 20 cellular oncogenes have been identified, and there is evidence that specific oncogenes are activated in particular human cancers. For example, an oncogene called ras is associated with many epithelial cancers, while another, called myc, is associated with leukemias.

An interesting feature of oncogenes is that they may act at different levels corresponding to the multiple steps seen in

Cancerous cell growth

the development of cancer. Some oncogenes immortalize cells, so that they divide indefinitely, whereas normal cells die after a limited number of generations. Other oncogenes transform cells so that they grow in the absence of growth factors. A combination of these two functions can make cells malignant, whereas each on its own cannot. The mode of action of oncogenes also provides important clues to the nature of growth control and cancer. At least some oncogenes are known to encode growth factors on the receptors of cells that may cause continuous proliferation in the absence of appropriate growth factors. Such studies identify the ways in which cancer cells differ from their normal counterparts, and they also emphasize how small and subtle those differences can be. (R.A.La.)

Cell differentiation

Adult organisms are composed of a number of distinct types of cell; in humans they number about 200. Cells are organized into tissues, each of which typically contains a small number of cell types and is devoted to a specific physiological function. For example, the epithelial tissue lining the stomach consists of three principal cell types: mucus-secreting columnar cells, acid-secreting parietal cells, and pepsin-secreting chief cells, together with smaller numbers of enteroendocrine cells secreting specific peptide hormones. The skeletal tissue of a long bone contains osteoblasts in the outer sheath; within the matrix are osteocytes, or mature bone cells, and osteoclasts, multinucleate cells involved in bone remodeling.

Invertebrate animals are also composed of differentiated cell types, many of which are similar to the cells of mammals; but, in general, the simpler the overall organization of the animal, the fewer the number of distinct cell types that they possess. Plants are also composed of differentiated cells, but they are quite different from the cells of animals. For example, a leaf in a higher plant is covered with a cuticle layer of epidermal cells. Among these are pores composed of two specialized cells, which regulate gaseous exchange across the epidermis. Within the leaf is the mesophyll, a spongy tissue responsible for photosynthetic activity. There are also veins composed of xylem elements, which transport water up from the soil, and phloem elements, which transport products of photosynthesis to the storage organs.

The various cell types have traditionally been recognized and classified according to their appearance in the light microscope following the process of fixing, processing, sectioning, and staining tissues that is known as histology. Classical histology has been augmented since the 1940s by electron microscopy, by histochemistry (methods of staining particular enzymes and other substances), and, since the 1960s, by immunohistochemistry (the use of specific antibodies to stain particular molecular species in situ). Immunohistochemistry has allowed the identification of many more cell types than could be visualized by classical histology, particularly in the immune system and among the scattered hormone-secreting cells of the endocrine system.

THE DIFFERENTIATED STATE

The biochemical basis of cell differentiation is the synthesis by the cell of a particular set of proteins, carbohydrates, and lipids. These syntheses are catalyzed by enzymes, all of them proteins, and each enzyme in turn is synthesized in accordance with a particular gene, or sequence of nucleotides in the DNA of the cell nucleus. A particular state of differentiation, then, corresponds to the set of genes that is expressed and the level to which it is expressed.

Genetic basis of differentiation

It is generally believed that all of an organism's genes are present in each cell nucleus, no matter what the cell type, and that differences between tissues are due not to the presence or absence of certain genes but to the expression of some and the repression of others. In animals the best evidence for retention of the entire set of genes comes from experiments in which the nucleus of a differentiated cell is substituted for the nucleus of a fertilized ovum, resulting in a small number of cases in the growth of a normal embryo. Such experiments show that any nucleus has the genetic information required for the growth of a developing organism, and they strongly suggest that, for most tissues, cell differentiation arises from the regulation of genetic activity rather than the removal or destruction of unwanted genes. The only known exception to this rule comes from the immune system, where segments of DNA in developing white blood cells are slightly rearranged, producing a wide variety of antibody and receptor molecules. (See above *The nucleus: Rearrangement and modification of DNA*.)

At the molecular level there are many ways in which the expression of a gene can be differentially regulated in different cell types. There may be differences in the copying, or transcription, of the gene into RNA, in the processing of the initial RNA transcript into messenger RNA (mRNA), in the control of mRNA movement to the cytoplasm, in the translation of mRNA to protein, or in the stability of mRNA. One of the best-understood cases is the differentiation of the red blood cell, of which hemoglobin is the major protein. In other tissues the globin genes are present but not transcribed at a high level, while, in the DNA of the developing red blood cell, nucleotide sequences called enhancers have been identified in and around the regions encoding hemoglobin. These enhancers, presumably upon interaction with regulatory proteins specific to red blood cells, allow the transcription of hemoglobin genes only in those cells.

Cells are normally sharply different from one another when they reach a state of terminal, or final, differentiation; intermediate forms found in an adult organism are in a process of maturation from a precursor cell type. States of terminal differentiation are stable and persistent, both in the lifetime of the cell and, in the case of differentiated types capable of continued cell division, through successive cell generations. The inherent stability of the differentiated state may be the result of a feedback mechanism, in which the products of active genes act on those genes to maintain their activity and on other genes to maintain their inactivity.

THE PROCESS OF DIFFERENTIATION

Differentiation from visibly undifferentiated precursor cells occurs during embryonic development from ova, during metamorphosis of larval forms, and following the separation of parts in asexual reproduction. It also takes place in the adult organism during the regeneration of missing parts and the growth of renewal tissues. Cell differentiation is, therefore, an essential and ongoing process at all stages of life.

Embryonic differentiation. The visible differentiation of cells is only the last of a progressive sequence of states, in each of which the cell becomes committed to a narrower range of types into which it can develop. Each committed state is probably based on differential gene activity, although this usually cannot be directly visualized. Different states of commitment can be investigated, however, by using the methods of experimental embryology. There are two basic procedures. First, the cell or group of cells is explanted and cultured in a neutral medium, and the types of cells into which it differentiates are recorded. This measures "specification," or the autonomous developmental capability of the cell. Second, the cell or group of cells is transplanted to other sites in the embryo in order to test its resistance to environmental stimuli. If it becomes the type of cell that it normally would regardless of position, it is said to be "determined" to form that particular tissue or cell type. Although states of specification and determination are thought to represent differential gene activity, the dynamic properties of embryonic cells are not necessarily the same as those of fully differentiated cells; in particular, they are often not stable over time.

Specification and determination of embryonic cells

Three mechanisms are thought to bring about altered commitments in the different regions of the early embryo: cytoplasmic localization, progressive chromosome modification, and induction. In the earliest stages of development it is usual for the ovum to divide without growth. The cells arising from these cleavage divisions are called blastomeres, and each blastomere inherits a certain region of the ovum. If a genetic regulatory molecule, called a cytoplasmic determinant, is localized to specific regions of the ovum, it will be inherited by only some of the blastomeres, providing a mechanism for the establishment of different states of commitment between blastomeres. Cytoplasmic determinants are thought to exist in most animal ova.

As mentioned above, chromosome modification as a mechanism of cell differentiation has been well established only in the immune system. However, it is thought to be a feature of early development in at least some animal types. The nematode *Caenorhabditis* has a cell lineage that does not vary between individual cells, but a number of mutations have been isolated in which the normal sequence of cell divisions and, hence, the resulting anatomy, are altered. The nature of these changes suggests that some form of chromosome modification is associated with particular blastomere cleavages, resulting in different determination of the daughter cells.

The role of induction in cellular commitment during embryonic development is very well established. In induction, a substance secreted by one group of cells alters the development of another group. In early development, induction is usually instructive; that is, the tissue assumes a different state of commitment in the presence of the signal than it would in the signal's absence. It is thought that inductive signals often take the form of concentration gradients of substances that evoke a number of different responses at different concentrations, leading to the formation of a sequence of groups of cells, each in a different state of specification.

In the early stages of vertebrate embryo development, the first induction, a signal from the endoderm acting on the cells of the ectoderm, leads to formation of the mesoderm. The second induction leads to a differentiation of mesoderm into somite (future vertebrae and muscle), lateral-plate, and blood-forming regions under the influence of the most dorsal region. During gastrulation, the third induction leads to the formation of the neural plate (the future brain and spinal cord) under the influence of the underlying mesoderm.

In later development, inductions may be merely permissive; that is, the signal allows completion of development along a certain pathway but does not increase cellular diversity. For example, in the development of the pancreas a factor from the mesenchymal tissue of the organ rudiment is known to be necessary for the proper differentiation of epithelial tissue.

Adult differentiation. In the adult mammal there are three classes of tissue. First, there are those that, like nerve and muscle tissue, are incapable of further cell division. Second, there are those that, like tissue of the liver or kidney, do not normally grow but may do so following removal of part of the tissue or after hormonal stimulation. Third, there are the renewal tissues, in which differentiated cells are constantly produced from an undifferentiated stem cell population.

Most renewal tissues are epithelia, such as the skin or the linings of the alimentary canal. But the renewal tissue whose differentiation is best understood is the blood. There are several different blood cell types: red blood cells, or erythrocytes, concerned with oxygen transport; granulocytes, mainly responsible for ingesting foreign particles; lymphocytes, concerned with antibody secretion and cell-mediated immune responses; and monocytes, responsible for removing cellular debris. There are also the platelets, fragments of giant cells called megakaryocytes, which are required for clotting. All of these types arise from a single type of stem cell, which is found in bone marrow together with larger numbers of mature cells and cells in intermediate stages of maturation.

Pluripotent stem cells

Evidence for a single, pluripotent stem cell (a stem cell from which many different types develop) comes mainly from experiments in which marrow cells are injected into mice whose blood cell production has been destroyed by irradiation. A small number of the injected cells grow in the spleen of the host, producing colonies, some of which contain more than one differentiated cell type. It can be shown that all the types in such colonies are descended from one stem cell. Experiments provide evidence about the events producing this differentiation: two cell types of restricted developmental potential arise immediately from the stem cell, one capable of forming granulocytes and monocytes and the other, erythrocytes and megakaryocytes. Each of these two types then segregates into two lines of unipotent cells (capable of producing only one type). Control over the number of terminal cell types occurs not at the level of cell commitment, which appears to be a random process, but by the action of growth-promoting substances on each of the lines at a precursor stage when they are committed but visibly undifferentiated and still capable of extensive cell division. Erythropoietin, for example, is required for the proliferation of erythroid precursors, and "granulocyte/macrophage colony-stimulating factor" is required for the proliferation of granulocyte/macrophage precursors.

Studies suggest that a similar situation obtains with epithelial renewal tissues but with an additional factor of spatial organization. For example, the small intestine is lined with crypts, each constituting one cell-proliferation unit. The crypts contain a small number of slowly cycling stem cells, which form all of the types of cells—absorptive, goblet, Paneth, and enteroendocrine—found in mucous membrane. These stem cells are located near the crypt base, and most of their progeny, destined to become absorptive and goblet cells, migrate upward into the villi. The intermediate forms between the stem cell and the differentiated cells go through several divisions, but little else is known about them. It is clear, however, that a proper balance between cell production and cell differentiation is necessary for the integrity and proper function of the tissue, for if the balance is disturbed, the tissue may become cancerous.

ERRORS IN DIFFERENTIATION

Two classes of abnormal cell differentiation are metaplasia and anaplasia. Metaplasia is the conversion of one cell type into another. In fact, it is not the differentiated cells themselves that change but rather the stem cell population from which they are derived. Metaplasia commonly occurs where chronic tissue damage is followed by extensive regeneration. Squamous metaplasia of the bronchi, for example, occurs when the ciliated respiratory epithelial cells of people who smoke develop into squamous, or flattened, cells. In intestinal metaplasia of the stomach, patches resembling intestinal tissue arise in the gastric mucosa, often in association with gastric ulcers.

Cancerous cells

Anaplasia is a loss of visible differentiation that can occur in advanced cancer. In general, early cancers resemble their tissue of origin, but as they develop they produce variants of more abnormal appearance and increased malignancy. Finally, a highly anaplastic growth can occur, in which the cancerous cells bear no visible relation to the parent tissue. The progression to anaplasia may represent a break in normal maturation, so that anaplastic cells resemble normal stem cells but do not respond to controls on growth and differentiation. (J.M.W.S.)

The evolution of cells

It is highly unlikely that scientists will ever re-create the crucial "experiment" that led to the origin of life. Billions of unsuccessful experiments must have been carried out in countless ponds and marshes before life first evolved, and these experiments lasted for hundreds of millions of years. During this period, conditions on Earth were different from those today. There was little oxygen in the atmosphere, and it is thought that the environment was rich in hydrogen, methane, and ammonia. Laboratory experiments show that, under these conditions, lightning charges could have created a large variety of small organic molecules, including the nucleotides, amino acids, and other raw materials of which cells are made. It is thought that life first arose after thick soups of such organic material had accumulated in local regions on Earth.

THE DEVELOPMENT OF GENETIC INFORMATION

Life could not exist until a collection of specific catalysts appeared that could promote the synthesis of more cat-

Primitive RNA

alysts of the same kind. Early stages in the evolutionary pathway presumably centred on RNA molecules, which not only present specific catalytic surfaces but also contain the potential for their own duplication through the formation of a complementary RNA molecule. It is assumed that a small RNA molecule eventually appeared that was able to catalyze its own duplication. Such an autocatalytic RNA molecule would have multiplied faster than its neighbours, usurping the RNA precursor molecules in the primeval soup.

Primitive RNA replication would have been imperfect, so that many variant autocatalytic RNA molecules would have arisen. Any variations that increased the speed or the fidelity of self-replication would have enabled those variant RNA molecules to outmultiply their neighbour RNA. Simultaneously, there would have been the natural selection of other small RNA molecules existing in symbiosis with autocatalytic RNA molecules, being replicated in return for catalyzing a useful secondary reaction such as the production of better precursor molecules. In this way, sophisticated families of RNA catalysts would eventually have evolved, in which cooperation between different molecules produced a system that was much more effective at self-replication than a collection of individual RNA catalysts.

The next major step in the evolution of the cell would have been the development, in one family of self-replicating RNA, of a primitive mechanism of protein synthesis. Protein molecules cannot provide the information for the synthesis of other protein molecules like themselves. This information must ultimately be derived from a nucleic acid sequence. Protein synthesis is much more complex than RNA synthesis, and it could not have arisen before a group of powerful RNA catalysts evolved. Each of these catalysts presumably has its counterpart among the RNA molecules that function in the current cell: (1) There was an information RNA molecule, much like messenger RNA (mRNA), whose nucleotide sequence was read to create an amino acid sequence; (2) there was a group of adaptor RNA molecules, much like transfer RNA (tRNA), that could bind to both mRNA and a specific activated amino acid; and (3) finally, there was an RNA catalyst, much like ribosomal RNA (rRNA), that facilitated the joining together of the amino acids aligned on the mRNA by the adaptor RNA.

Formation of cell membranes

At some point in the evolution of biologic catalysts the first cell was formed. This would have required the partitioning of the primitive soup of biologic catalysts into individual units, each surrounded by a membrane. Membrane formation might have occurred quite simply, since many amphiphilic molecules—half hydrophobic (water-hating) and half hydrophilic (water-loving)—aggregate to form bilayer sheets in which the hydrophobic portions of the molecules line up in rows to form the interior of the sheet and leave the hydrophilic portions to face the water. Such bilayer sheets can spontaneously close up to form the walls of small, spherical vesicles, as do the phospholipid bilayer membranes of present-day cells.

As soon as the biologic catalysts became compartmentalized into small individual units, or cells, the units would have begun to compete with one another for the same ingredients in the surrounding soup. Now the development of variant, but efficient, catalysts would have served only the cell itself and its progeny, rather than being dissipated throughout a much larger volume. The active competition that ensued must have greatly speeded up evolutionary change, serving as a powerful force for the development of more efficient cells. In this way, cells eventually arose that contained new catalysts, enabling them to use simpler, more abundant precursor molecules for their growth. Because these cells were no longer dependent on a rich soup of preformed ingredients for their survival, they were able to spread far beyond the limited environments where the first primitive cells arose.

It is often assumed that the first cells appeared only after the development of a primitive form of protein synthesis. However, it is by no means certain that cells cannot exist without proteins, and it has been suggested as an alternative that the first cells contained only RNA catalysts. In either case, protein molecules, with their chemically varied side chains, are more powerful catalysts than RNA molecules; therefore, as time passed, cells arose in which RNA served primarily as genetic material, being directly replicated in each generation and inherited by all progeny cells in order to specify proteins.

As cells became more complex, a need would have arisen for a stabler form of genetic information storage than that provided by RNA. DNA, related to RNA yet chemically stabler, probably appeared rather late in the evolutionary history of cells. Over a period of time, the genetic information in RNA sequences was transferred to DNA sequences, and the ability of RNA molecules to replicate directly was lost. It was only at this point that the central process of biology—the synthesis, one after the other, of DNA, RNA, and protein—appeared.

THE DEVELOPMENT OF METABOLISM

Transition from anaerobic to aerobic cells

The first cells presumably resembled prokaryotic cells in lacking nuclei and functional internal compartments, or organelles. These early cells were also anaerobic (not requiring oxygen), deriving their energy from the fermentation of organic molecules that had previously accumulated on the Earth over long periods of time. As these food sources became exhausted, more sophisticated cells evolved that could carry out primitive forms of photosynthesis, in which light energy was harnessed by membrane-bound proteins to form organic molecules with energy-rich chemical bonds. A major turning point in the evolution of life was the development of photosynthesizing bacteria requiring only water as an electron donor and producing molecular oxygen. The descendants of these bacteria, the cyanobacteria (also known misleadingly as blue-green algae), still exist as viable life forms. Their ancestors prospered to such an extent that the atmosphere became rich in the oxygen they produced. The free availability of this oxygen in turn enabled other bacteria to evolve aerobic forms of metabolism, a much more efficient use of organic molecules as a source of food.

The switch to predominantly aerobic metabolism is thought to have occurred in bacteria approximately 2,000,000,000 years ago, about 1,500,000,000 years after the first cells had formed. Aerobic eukaryotic cells (cells containing nuclei and all of the other organelles) probably appeared about 1,500,000,000 years ago, their lineage having branched off much earlier from that of the bacteria. Eukaryotic cells almost certainly became aerobic by engulfing aerobic bacteria, with which they lived in a symbiotic relationship. The mitochondria found in both animals and plants are the descendants of such bacteria. Later, in branches of the eukaryotic lineage leading to plants and algae, a cyanobacterium-like organism was engulfed to perform photosynthesis; in time, these bacteria became the chloroplasts.

The eukaryotic cell thus apparently arose as an amalgam of different cells, in the process becoming an efficient aerobic cell whose plasma membrane was freed from energy metabolism—one of the major functions of the plasma membrane of bacteria. The eukaryotic plasma membrane was therefore able to become specialized for cell-to-cell communication and cell signaling. It may be partly for this reason that eukaryotic cells were eventually more successful at forming complex multicellular organisms than their simpler prokaryotic relatives. (B.M.A.)

The history of cell theory

FORMULATION OF THE THEORY

Early observations. The history of cell theory is a history of the actual observation of cells, because early prediction and speculation about the nature of the cell were generally unsuccessful. The decisive event that allowed the observation of cells was the invention of the microscope in the 17th century, after which interest in the invisible world was stimulated. Robert Hooke, who described cork and other plant tissues in 1665, introduced the term cell because the cellulose walls of dead cork cells reminded him of the blocks of cells occupied by monks. Even after the publication in 1672 of excellent pictures of plant tis-

The role of microscopy

sues, no significance was attached to the contents within the cell walls. The magnifying powers of the microscope and the inadequacy of techniques for preparing cells for observation precluded a study of the intimate details of the cell contents. The inspired amateur of early microscopy Antonie van Leeuwenhoek, beginning in 1673, discovered blood cells, spermatozoa, and a lively world of "animalcules." A new world of unicellular organisms was opened up. Such discoveries extended the known variety of living things but did not bring insight into their basic uniformity. Moreover, when Leeuwenhoek observed the swarming of his animalcules but failed to observe their division, he could only reinforce the idea that they arose spontaneously.

Cell theory was not formulated for nearly 200 years after the introduction of microscopy. Explanations for this delay range from the poor quality of the microscopes to the persistence of ancient ideas concerning the definition of a fundamental living unit. Many observations of cells were made, but apparently none of the observers was able to assert forcefully that cells were the units of biologic structure and function.

Three critical discoveries made during the 1830s, when improved microscopes with suitable lenses, higher powers of magnification without aberration, and more satisfactory illumination became available, were decisive events in the early development of cell theory. First, the nucleus was observed by Robert Brown in 1833 as a constant component of plant cells. Next, nuclei were also observed and recognized as such in some animal cells. Finally, a living substance called protoplasm was recognized within cells, its vitality made evident by its active streaming, or flowing, movements, especially in plant cells. After these three discoveries, cells, previously considered as mere pores in plant tissue, could no longer be thought of as empty, because they contained living material.

Two German biologists, Theodore Schwann and Matthias Schleiden, clearly stated in 1839 that cells are the "elementary particles of organisms" in both plants and animals and recognized that some organisms are unicellular and others multicellular. This statement was made in Schwann's *Mikroskopische Untersuchungen über die Übereinstimmung in der Struktur und dem Wachstume der Tiere und Pflanzen* (1839; *Microscopical Researches into the Accordance in the Structure and Growth of Animals and Plants*). Schleiden's contributions on plants were acknowledged by Schwann as the basis for his comparison of animal and plant structure.

Schleiden and Schwann's descriptive statements concerning the cellular basis of biologic structure are straightforward and acceptable to modern thought. They recognized the common features of cells to be membrane, nucleus, and cell body and described them in comparisons of various animal and plant tissues. A statement by Schleiden pointed toward the future direction of cell studies:

> Each cell leads a double life: an independent one, pertaining to its own development alone; and another incidental, insofar as it has become an integral part of a plant. It is, however, easy to perceive that the vital process of the individual cells must form the first, absolutely indispensable fundamental basis, both as regards vegetable physiology and comparative physiology in general. . . .

The problem of the origin of cells. Schwann and Schleiden were not alone in contributing to this great generalization of natural science, for strong intimations of the cell theory occur in the work of their predecessors. Recognizing that the basic problem was the origin of cells, these early investigators invented a hypothesis of "free cell formation," according to which cells developed *de novo* out of an unformed substance, a "cytoblastema," by a sequence of events in which first the nucleolus develops, followed by the nucleus, the cell body, and finally the cell membrane. The best physical model of the generation of formed bodies then available was crystallization, and their theory was inspired by that model. In retrospect, the hypothesis of free cell formation would not seem to have been justified, however, since cell division, a feature not characteristic of crystallization processes, had frequently been observed by earlier microscopists, especially among single-celled organisms. Even though cell division was observed repeatedly in the following decades, the theory of free cell formation lingered throughout most of the 19th century; however, it came to be thought of more and more as a possible exception to the general principle of the reproduction of cells by division. The correct general principle was affirmed in 1855 by a German biologist of great prestige, Rudolph Virchow, who asserted that "omnis cellula e cellula" ("all cells come from cells"). Doubt remained, however.

The hypothesis of "free cell formation"

The inherently complex events of cell division prevented a quick resolution of the complete sequence of changes that occur during the process. First, it was noted that a cell with a nucleus divides into two cells, each having a nucleus; hence, it was concluded that the nucleus must divide, and direct division of nuclei was duly described by some. Better techniques served to create perplexity, because it was found that, during cell division, the nucleus as such disappears. Moreover, at the time of division, dimly discerned masses, now recognized as chromosomes, were seen to appear temporarily. Observations in the 1870s culminated in the highly accurate description and interpretation of cell division by Walther Flemming in 1882. His advanced techniques of fixing and staining cells enabled him to see that cell reproduction involves the transmission of chromosomes from the parent to daughter cells by the process of mitosis and that the division of the cell body is the terminal event of that reproduction.

The discovery that the number of chromosomes remains constant from one generation to the next resulted in the full description of the process of meiosis. The description of meiosis, combined with the observation that fertilization is fundamentally the union of maternal and paternal sets of chromosomes, culminated in the understanding of the physical basis of reproduction and heredity. Meiosis and fertilization therefore came to be understood as the complementary events in the life cycle of organisms: meiosis halves the number of chromosomes in the formation of spores (plants) or gametes (animals), while fertilization restores the number through the union of gametes. By the 1890s "life" in all of its manifestations could be thought of as an expression of cells.

Protoplasm as the living substance of the cell

The protoplasm concept. As the concept of the cell as the elementary particle of life developed during the 19th century, it was paralleled by the "protoplasm" concept—the idea that the protoplasm within the cell is responsible for life. Protoplasm had been defined in 1835 as the ground substance of living material and hence responsible for all living processes. That life is an activity of an elementary particle, the cell, can be contrasted with the view that it is the expression of a living complex substance—even a supermolecule—called a protoplasm. The protoplasm concept was supported by observations of the streaming movements of the apparently slimy contents of living cells.

Advocates of the protoplasm concept implied that cells were either fragments or containers of protoplasm. Suspicious and often contemptuous of information obtained from dead and stained cells, such researchers discovered most of the basic information on the physical properties—mechanical, optical, electrical, and contractile—of the living cell.

An assessment of the usefulness of the concept of protoplasm is difficult. It was not wholly false; on the one hand, it encouraged the study of the chemical and mechanical properties of cell contents, but it also generated a resistance, evident as late as the 1930s, to the development of biochemical techniques for cell fractionation and to the realization that very large molecules (macromolecules) are important cellular constituents. As the cell has become fractionated into its component parts, protoplasm, as a term, no longer has meaning. The word protoplasm is still used, however, in describing the phenomenon of protoplasmic streaming—the phenomenon from which the concept of protoplasm originally emerged.

CONTRIBUTION OF OTHER SCIENCES

Appreciation of the cell as the unit of life has accrued from important sources other than microscopy; perhaps

The role of microbiology

the most important is microbiology. Even though the small size of microorganisms prohibited much observation of their detailed structure until the advent of electron microscopy, they could be grown easily and rapidly. Thus it was that the studies of microbes by Louis Pasteur, published in 1861, helped to establish the principle of biogenesis, namely, that organisms arise only by the reproduction of other organisms. Fundamental ideas regarding the metabolic attributes of cells—that is, their ability to transform simple nutritional substances into cell substance and utilizable energy—came from microbiology. Pasteur perhaps overplayed the relation between catalysis and the living state of cells in considering enzymatic action to be an attribute of the living cell rather than of the catalytic molecules (enzymes) contained in the cell; it is a fact, however, that much of cell chemistry is enzyme chemistry—and that enzymes are one defining attribute of cells. The techniques of microbiology eventually opened the way for microbial genetics, which, in turn, provided the means for solving the fundamental problems of molecular biology that were inaccessible at first to direct attack by biochemical methods.

The science of molecular biology would be most capable of overthrowing the cell theory if the latter were an exaggerated generalization. On the contrary, molecular biology has become the foundation of cell science, for it has demonstrated not only that basic processes such as the genetic code and protein synthesis are similar in all living systems but also that they are made possible by the same cell components—*e.g.,* chromosomes, ribosomes, and membranes.

In the overlapping histories of cell biology and medicine, two events are especially important. One, the identification in 1827 by Karl Ernst von Baer of the ovum (unfertilized egg) as a cell, was important considering the many ways it often differs from other cells. Baer not only laid the foundations for reproductive biology but also provided important evidence for the cell theory at a critical time. The second important event was the promotion in 1855 of the concept of "cellular pathology" by Rudolph Virchow. His idea that human diseases are diseases of cells and can be identified and understood as such gave an authority to cell theory.

Although biochemistry might have made considerable progress without cell theory, each influenced the other almost from the start. When it was established that most biochemical phenomena are shared by all cells, the cell could be defined by its metabolism as well as by its structure. Cytochemistry, or histochemistry, made a brilliant start in 1869, when a Swiss biochemist, Friedrich Miescher, postulated that the nucleus must have a characteristic chemistry and then went on to discover nucleic acids, which have since been shown to be the crucial molecules of inheritance and metabolism.

The cell as part of an organism

Cell theory by itself cannot explain the development and unity of the multicellular organism. A cell is not necessarily an independently functioning unit, and a plant or an animal is not merely a congeries of individual cells. Fortunately, however, the long controversy centring upon the individuality and separateness of cells has ended. Cell biology now focuses on the interactions and communication among cells as well as on the analysis of the single cell. The influence of the environment on the cell has always been considered important; now it has been recognized that one important part of the environment of a cell is other cells.

Cell theory thus is not so comprehensive as to eliminate the concept of the organism as more than the sum of its parts. But the study of a particular organism requires the investigation of cells as both individuals and groups. The problem of cancer is an example: a plant or animal governs the division of its own cells; the right cells must divide, be differentiated, and then integrated into the proper organ system at the right time and place. Breakdown results in a variety of abnormalities, one of which is cancer. When the cell biologist studies the problem of the regulation of cell division, the ultimate objective is to understand the effect of the whole organism on an individual cell.

(B.M.A./Ed.)

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(B.M.A./W.D.S./R.A.La./M.R.B./L.A.St./J.M.W.S.)

Celtic Literature

This article gives a historical account of Celtic literature, a term commonly applied to writings in Irish Gaelic, Scottish Gaelic, Manx, Welsh, Cornish, and Breton, in the so-called Insular Celtic languages. For writings in English by Irish, Scottish, and Welsh authors, see ENGLISH LITERATURE. French-language works by Breton authors are covered in FRENCH LITERATURE.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* section 621, and the *Index.*

The article is divided into the following sections:

IRISH GAELIC

The introduction of Celtic into Ireland has not been authoritatively dated, but it cannot be later than the arrival there of the first settlers of the La Tène culture in the 3rd century BC. The language is often described in its earliest form as Goídelic, named after the Celts (Goídil; singular, Goídel) who spoke it. The modern form, known in English as Gaelic (in Gaelic called Gaedhilge or Gaeilge), is derived from the Scottish Gàidhlig.

The earliest evidence of Irish Gaelic consists of archaic sepulchral inscriptions in the ogham alphabet based on a system of strokes and notches cut on the edges of stone or wood usually ascribed to the 4th and 5th centuries AD. Writings in the Roman alphabet date from 8th-century glosses in Old Irish, but 7th- and even 6th-century compositions are preserved in much later manuscripts.

Four distinct periods are recognizable in Irish Gaelic literature. The early literature (linguistically Archaic, Old, and Early Middle Irish), was composed by a professional class, the fili (singular, fili), and by churchmen. The medieval literature (linguistically late Middle and Classical Modern Irish) was dominated by the lay and hereditary bardic orders. In the late literature (17th century to the end of the 19th) authorship passed into the hands of individuals among the peasants, the class to which most Irish speakers had been reduced, using the dialects into which the language had been broken up. The subsequent revival has continued to the present day.

The functions of the fili

Early period. Irish literature was originally aristocratic and was cultivated by the fili, who seem to have inherited the role of the learned priestly order represented in Caesar's Gaul by the Druids, *vates* ("seers"), and bards and to have been judges, historians, and official poets responsible for all traditional lore and the performance of all rites and ceremonies. The arrival of Christianity and the gradual disappearance of paganism led to the abandonment of their specifically priestly functions. Nevertheless, the fili seem to have retained responsibility for the oral transmission of native lore or learning, which was in marked contrast to the new book or manuscript learning of the Christian Church. Fortunately, the ecclesiastical scholars were not as hostile to the native lore as were their counterparts abroad, and they appear to have been eager to commit it to writing. As a result, Ireland's oral culture was extensively recorded in writing long before it could have evolved that art itself. The record consisted mainly of history, legendary and factual; laws; genealogies; and poems, but prose was the predominant vehicle.

The fili were powerful in early Irish society and were often arrogant, enforcing their demands by the threat of a lampoon (*áer*), a poet's curse that could ruin reputations and, so it was thought, even kill. The laws set out penalties for abuse of the *áer,* and belief in its powers continued up to modern times. The official work of the fili has been preserved in fragments of annals and treatises.

Rhyme and metre

Verse. The earliest verse has been preserved mainly in passages incorporated into later documents, both literary and legal; most have suffered in transmission and are very obscure. One of the earliest poems is a eulogy on St. Columba (*c.* 521–597) in rhetorical short sentences linked by alliteration, ascribed to Dallán Forgaill, chief poet of Ireland. This device of alliterative rhythmical prose was used again in the sagas. Probably the oldest actual metre was that in which two half lines were linked by alliteration—a system reminiscent of early Germanic verse. Rhyme was used from the 7th century; the requirement was only that there should be identity of vowel and syllabic length and that consonants should belong to the same phonetic class—a system also found in early Welsh. The quatrain (seven or eight syllables to a line and rhyme between second and fourth lines) was derived from Latin hymn metres. The quatrains of the later popular metre, the *debide* (literally "cut in two"), consisted of two couplets with the two lines of each couplet rhyming.

Much early verse was of an official nature, but that of the church was hardly more lively than that of the fili, who often affected a deliberately obscure style. More interesting was the 10th-century *Psalter,* a biblical history in 150 poems. But the real glory of Irish verse lay in anonymous poets who composed poems such as the famous address to Pangur, a white cat. They avoided complicated metres and used a language that had been cultivated for centuries, with a freshness of insight denied to the fili. That the fili could, however, adapt their technique was shown by an 11th-century poem on the sea, where preface, choice of theme, and metaphorical expressions all suggest Scandinavian influence. This and other nature poetry carried on a tradition of native lyrics, sagas, and seasonal songs that showed remarkable sensitivity. The monastic and eremitic movement in the Irish Church also provided a strong impetus to nature poetry. This almost Franciscan poetry had an especial appeal to monastic scribes, so that much of it has been preserved.

Historical verse arose partly because recording of the past was an important part of the work of the fili; some of the earliest poems were metrical genealogy. As time went on the necessity for compendiums of information grew, and these were again often in metrical form. In a long poem, *Fianna bátar in Emain* ("The Warriors Who Were in Emain"), Cináed ua Artacáin summed up the saga material, while Fland Mainistrech collected the work of generations of fili who had laboured to synchronize Ireland's history with that of the outside world. Equally important is a great collection, in prose and verse, called the *Dindshenchas,* which gave appropriate legends to famous sites of Ireland between the 9th and 11th centuries. Indeed, the development of a loose *debide* form, making rhyming easy, facilitated mnemonic verse on numerous subjects.

The early Irish epic

Prose. The early Irish epic was a prose narrative that usually contained non-narrative poetic passages, often in dialogue form. The resemblance between this and the type of epic found in early Sanskrit suggests that the tradition went back to Indo-European times. The oldest sagas probably were first written down in the 7th and 8th centuries, from an oral tradition. These were imperfectly preserved, since Scandinavian invasions in the 9th and 10th centuries disrupted literary studies. Not until the 11th century did life become sufficiently settled for works to be collected in monasteries, and even then the collecting seems to have been haphazard. The great codex *Lebor Na Huidre* (*The Book of the Dun Cow*), written early in the 12th century, showed older treatments of saga material than are found in *The Book of Leinster,* written years later. The material has preserved a picture of primitive society—fighting from chariots, taking heads as trophies, the position of the Druids, the force of taboos—for which there is little or no evidence from strictly historical sources.

The Ulster cycle

The most important cycle was that of the Ulaid, a people who gave their name to Ulster. Conchobar (king of the Ulaid), Cú Chulainn (a boy warrior), Medb (queen of Connaught), and Noísi and Deirdre, doomed lovers, were outstanding figures in early Irish literature, and it was on elements from the sagas of the Ulaid that the nearest approach to an Irish epic was built—*Táin Bó Cuailnge* (*The Cattle Raid of Cooley*). First put together in the 7th or 8th century, it is striking chiefly for its terse vividness. The finest section is that in which Fergus, an exile from Ulster, recalls the deeds of Cú Chulainn's youth. But the value of the *Táin* generally lies in that, as it was being continually worked over, it provides a record of the degeneration of Irish style. The *Táin* collected around itself a number of ancillary stories, including that of its revelation to the fili in the 7th century by the ghost of Fergus and that of the tragedy of Noísi and Deirdre.

In this period stories with an origin outside the recognized tradition began to appear. An exotic element was represented by *The Taking of Troy* and *The Story of Alexander,* which appeared in the oldest saga lists, but classical learning had comparatively little effect until the next period. Stories of Finn, whose traditions went back to an early period, only really developed when the fili were no longer in control. The "wild-man-of-the-woods" cycle associated with Suibne Geilt had its origins in Strathclyde, where Irish and Brythonic literature must have been in contact at an early date; this mixture of hagiography, saga, and nature material was one of the most attractive stories of the later period.

The earliest didactic writings, Irish in language but in content mostly of Latin origin, comprised monastic rules, homilies, and hagiography. The lives of the saints were mainly works of fantasy, increasingly incorporating elements from folklore and saga material. The emphasis was always on the miraculous, but they are valuable as social documents. Another important genre of religious work was the vision, exemplified in *Fís Adamnáin* (*The Vision of Adamnan*), whose soul is represented as leaving his body for a time to visit heaven and hell under the guidance of an angel. Both the saints' lives and the visions tended to degenerate into extravagance, so that parodies were composed, notably *Aislinge Meic Con Glinne* (*The Vision of MacConglinne*).

Theory has always been important in Irish cultural life, and the fili built up a considerable body of academic speculation. A few surviving fragments discuss the nature of inspiration and the origin of language, with practical instruction on matters of metrics and style. Early on, a technical vocabulary was built up to deal with Latin, as well as Irish, grammar. In several ancient texts discussion of the art of the poet was mixed with questions relating to his legal status.

The rise of the bards

Middle period. The 12th century was a period of contradictions. While the oldest surviving codices were being written in monasteries, fairly faithfully preserving the much earlier literature, a new literary order elaborated verse forms to a far greater extent than before and used a language close to the vernacular of the day. After the 12th century the hereditary bardic families became custodians of Irish literature and continued in that function until the collapse of the Gaelic polity. At first completely divorced from ecclesiastical influence, they soon were sending their sons into the recently introduced orders, above all to the Franciscans, who were to become by the end of this period the greatest custodians of the tradition.

Bardic verse. The bardic reform of verse was sweeping. The language was modernized, the large number of metres used by the fili was greatly reduced, and greater rhythmic control and ornamentation were required. The scope of the verse narrowed; the bulk was poetry praising the poet's patron or God. No longer associated with monastic foundations, the bards, who trained for six or seven years, confidently looked to patrons to secure their living. One of the earliest poets of the great bardic family of Ó Dálaigh, Muireadhach Albanach, left a fine elegy on the death of his wife, as well as a stirring defense of his action in killing a tax collector. The courtly love themes, introduced into Irish literature by the Norman invaders, were used with native bardic wit and felicitous style to produce the enchanting poems called *dánta grádha.* A different departure from praise poetry was the *crosánacht,* in which verse was frequently interspersed with humorous or satirical prose passages.

The Fenian cycle. Most native prose of this period was concerned with the hero Finn and his war band (*fian*). Stories about Finn, Oisín, Caoilte, and the rest must have existed among the people for many centuries. The outstanding work was *Agallamh na Seanórach* ("The Interrogation of the Old Men"), written in the 12th century, in which Caoilte is represented as surviving the Battle of Gabhra and living on to accompany Patrick through Ireland. The Fenian stories never received such careful literary treatment as did those of the Ulster cycle, and the old form was soon abandoned for prose tales and ballads, which may be regarded as the beginnings of popular, as opposed to professional, literature in Irish. The metres represented a drastic simplification of the bardic technique, and a distinct change in theme occurred as this literature passed into the hands of the people.

Other prose. Stories popular with the fili steadily dropped out of favour. Sometimes they were combined with folktale elements, as was the case with the very old saga of Fergus mac Léti, which was rewritten, perhaps in the 14th century, to include a story of a people of tiny stature—the leprechauns. Most important of all, a flood of translations from Latin and English began. The stories of Marco Polo, Sir John Mandeville, Prester John, and Guy of Warwick, as well as classical and Arthurian stories in their medieval adaptations, became well known in Ireland. The new religious orders translated many spiritual and devotional works, and the churchmen made the experiment, remarkable for the time, of handling philosophical material in the vernacular. There was also much technical writing, especially on grammar and metrics. Continental teaching seems to have superseded the native tradition during this period.

The influence of other literatures

The spread of printing

By the end of the 15th century the printing press began to make literature available to larger numbers in most European countries. In Ireland, however, literature remained for some time the preserve of those who could afford to maintain the writers and supply their costly vellum.

Late period. The dispossession of the Irish and the old Anglo-Irish nobility during the late 16th and early 17th centuries entailed the practical disappearance of the professional bards, who were the nobility's dependents and propagandists. With their elimination the old order was doomed, and the Irish language itself began its long process of decay.

Hardly any correct bardic verse was written in Ireland after 1650, but new poets took over from the bards. And just as the bardic measures had been in preparation for centuries before they established themselves as canonical, so the song metres that replaced them had existed for centuries among the people. The new poets abandoned the syllable-measured lines for lines with a fixed number of stresses; the stressed vowels rhymed in patterns that might be very simple or, later, bewilderingly intricate, but simple vocalic assonance took the place of earlier rhyme. The language of poetry moved toward that of the people. While poets had little patronage, there was at least an increasing supply of paper, so that their works, still barred from the printing press, were able to circulate. The tone of verse throughout the 17th century was passionately defiant of the new regime. In it is found the first coherent expression of patriotism conceived as devotion to an abstract ideal rather than as loyalty to an individual, but much of the verse represents a mere nostalgia for the past.

The greatest poets of the song metres were Dáibhidh Ó Bruadair, one of the last poets to enjoy some patronage, and Aodhagán Ó Rathoille whose aisling (vision) poems made the genre popular. After them the poetic tradition was maintained into the 19th century by peasant poets who, although not lacking in subtlety of craftsmanship, and occasionally vigorous in satire, had none of the advantages and only a few of the virtues of their predecessors.

During the 17th century valuable antiquarian prose was produced. The most important is *Annála Ríoghachta Éireann* (completed 1636; "Annals of the Kingdom of Ireland"; Eng. trans., *Annals of the Four Masters*), a compilation of all available material on the history of Ireland to 1616, directed by Michael O'Clery. Geoffrey Keating produced the first historical (as opposed to annalistic) work in his *Foras Feasa ar Éirinn* (written *c.* 1640; *History of Ireland*) as well as some fine verse in both old and new metres and two spiritual treatises.

Prose developments

An interesting development in prose style was the satire *Páirliment Chloinne Tomáis* ("Parliament of Clan Thomas"). It appears to be by a representative of the bardic order, for it attacked with equal savagery the new ruling class and the native peasantry, using a style close to that of the earlier *crosánacht* but with prose predominating over verse. It found several imitators, but the old tradition was by this time too attenuated for so aristocratic an attitude to be maintained. Imaginative prose was more popular; it consisted of developments of Fenian or romance themes from Irish and foreign medieval literature mingled with elements of folklore and of the fabliau (a short metrical tale). As in the case of the song metres, these romances had a considerable tradition before they appeared in writing. But as the public for Irish became smaller, there was little hope for much prose production.

The 18th century is a low point in Irish Gaelic literature. The last great flowering of the poetic tradition in Munster was *Cúirt an Mheadhon Oidhche* (written 1780, published 1904; *The Midnight Court*) by Brian Merriman, a Clare schoolmaster. After it, Irish poetry became a matter of folk songs.

During the 18th and early 19th centuries the only books in Irish prose were catechisms and devotional tracts. The manuscript tradition was carried on by a few scribes into the first half of the 19th century, when it all but died out. By the mid-19th century there was little literary activity in Irish, and almost all Irish speakers were illiterate.

The Gaelic revival. Ironically, it was English-speaking antiquarians and nationalists from the small educated class, rather than the Irish-speaking minority, who led the 19th-century revival, which in turn was stimulated by the Romantic movement's interest in Celtic subjects.

The rich vocabulary and idiomatic expressions and the wealth of folklore and folktales of the Irish-speaking districts (*gaeltachts*) gradually were acknowledged. Folklore collectors such as Douglas Hyde were able to restore some sense of pride in the language. The revivalists even succeeded in securing for Irish a modest place in the country's educational system.

But the revivalists were faced with a language of diverse dialects, and standardization was only effected in the mid-20th century with the help of new grammars, adequate dictionaries, and government support and direction. Writers whose Irish was rich and vigorous were persuaded that the reading public needed not more folklore but a literature that could compete internationally. Among the pioneers in this field were Patrick Pearse and Pádraic Ó Conaire, who introduced the modern short story into Irish. The short story flourished in the hands of Liam O'Flaherty and Máirtín O Cadhain, who also produced an outstanding novel, *Cré na Cille* (1953; "Churchyard Dust"). In verse the work of Máire Mhac an tSaoi, Máirtín Ó Direáin, and Seán Ó Ríordáin is impressive, and in their shadow a new generation of young poets, some of considerable talent, has grown up. Indeed, there is no lack of literary activity, as the two literary periodicals, *Comhar* ("Cooperation") and *Feasta* ("Henceforth"), eloquently testify.

In drama Brendan Behan's *An Giall* (1957; *The Hostage*) stands out, but Seán Ó Tuama, Eoghan Ó Tuairisc, Diarmaid Ó Súilleabháin, and Criostóir Ó Floinn also have written fine plays and have contributed in other genres. However, the drama in Irish cannot be said to have created for itself the following attracted by the Anglo-Irish theatre in its heyday under the influence of Yeats and Synge.

Literary criticism of distinction has developed also, and Seán Ó Tuama, Tomás Ó Floinn, and Breandán Ó Doibhlin have produced essays of high standard.

The most valuable contribution made by the *gaeltachts* has been a series of personal reminiscences describing local life. One of the best is Tomás Ó Criomhthain's *An tOileánach* (1929; *The Islandman*). At one time the *gaeltacht* memoirs threatened to become a vogue and inspired the brilliant satirical piece *An Béal Bocht* (1941; *The Poor Mouth*) by Flann O'Brien (pseudonym of Brian Ó Nualláin). Less characteristic but perhaps no less valuable have been the autobiographies written in Irish. Together with the spate of scholarly biographies in Irish, some on literary or semiliterary figures, they show how the revival has increased its range and depth.

SCOTTISH GAELIC

The Lismore compilation

Writings of the medieval period. The earliest extant Scottish Gaelic writing consists of marginalia added in the 12th century to the Latin Gospels contained in the 9th-century *Book of Deer.* The most important early Gaelic literary manuscript is *The Book of the Dean of Lismore,* an anthology of verse compiled between 1512 and 1526 by Sir James MacGregor, dean of Lismore (Argyllshire), and his brother Duncan. Its poems fall into three main groups: those by Scottish authors, those by Irish authors, and ballads concerned with Ossian, the mythical warrior and bard. This is the earliest extensive anthology of heroic Gaelic ballads in either Scotland or Ireland. The Scottish Gaelic poems date from about 1310 to 1520. The bard best represented is Fionnlagh Ruadh, bard to John, chief of clan Gregor (died 1519). There are three poems by Giolla Coluim mac an Ollaimh, a professional poet at the court of the Lord of the Isles and almost certainly a member of the MacMhuirich bardic family, the famous line of hereditary bards whose work spans nearly 500 years from the 13th to the 18th century. Perhaps the most notable of the other poets is Giolla Críost Brúilingeach and two women, Aithbhreac Inghean Coirceadail and Isabella, countess of Argyll.

Continuation of the oral tradition. Some 16th-century Gaelic poetry survived in oral tradition until the mid-18th century, when it was written down. Examples are *An Duanag Ullamh* ("The Finished Poem"), composed in honour of Archibald Campbell, 4th earl of Argyll, and the lovely lament *Griogal Cridhe* ("Teasing Heart"; *c.* 1570). It is certain that the poetry recorded in *The Book of the Dean of Lismore* was not an isolated outburst; much professional and popular poetry must have been lost. Songs in the nonsyllabic, accented measures survived, again orally, from the early 17th century. This was the tradition that produced the work songs—*e.g.,* waulking songs used when fulling cloth.

In 1567 appeared the first book printed in Gaelic in Scotland: Bishop John Carswell's *Foirm na Nurrnuidheadh,* a translation of John Knox's liturgy, in Classical Common Gaelic.

The 17th century. This period was the high point of Scottish Gaelic literature. The political, ecclesiastical, and social structures of Scotland were changing as was the relationship between the central government and the Gaelic area. Enough Gaelic poetry survives to show that there were many poets of great talent, and the diffusion of artistic talent is scarcely matched in any other period in Scottish Gaelic history. It was the great age of the work songs and of the classical bagpipe music. Some of the poetry and prose was contained in three 17th-century manuscripts. The first two were the *Black Book of Clanranald* and the *Red Book of Clanranald,* written by members of the MacMhuirich family, who were latterly hereditary bards to the MacDonalds of Clanranald. They were probably written for the most part in the 17th century but contained poems by earlier representatives of the family. The other important document was the Fernaig manuscript, compiled between 1688 and 1693, containing about 4,200 lines of verse, mostly political and religious.

Mary MacLeod and Iain Lom

The two best known poets of the 17th century were Mary MacLeod and Iain Lom. The former, known as Màiri Nighean Alasdair Ruaidh, was closely associated with the MacLeods of Harris and Dunvegan. Her poems show deep personal emotion, and her style is fresh and natural. She inherited the imagery of the bardic poets but placed it in a new setting, and her metres were strophic (having repeating patterns of lines) rather than strictly syllabic. John Macdonald, known as Iain Lom, took an active part in the events of his time. His life spanned an eventful period in Highland history, and his poetry reflected this. He composed poems about the battles of Inverlochy and Killiecrankie, a lament for the Marquess of Montrose, a poem on the Restoration of Charles II in 1660, several poems dealing with the Keppoch murder of 1663, and a song bitterly opposing the union of the Parliaments in 1707. His versification had a compression and concentration not often found in later Gaelic poetry.

Other noteworthy 17th-century poets include Donnchadh MacRaoiridh, whose best-known poem consists of four calm, resigned verses composed on the day of his death; Alasdair MacKenzie and his son Murdo Mackenzie; and Roderick Morison, known as An Clarsair Dall (the Blind Harper), who became harper to Iain Breac MacLeod of Dunvegan. The strong texture and poetic intensity of Morison's *Oran do Iain Breac MacLeòid* ("Song to Iain MacLeod") and his *Creach na Ciadaoin* ("Wednesday's Bereavement") are remarkable. Dorothy Brown and Sìleas na Ceapaich were women poets of great talent.

Four other poets mark the transition from the poetry of the 17th century to that of the 18th: Lachlan MacKinnon (Lachlann Mac Thearlaich Oig); John Mackay (Am Pìobaire Dall), whose *Coire an Easa* ("The Waterfall Corrie") was significant in the development of Gaelic nature poetry; John Macdonald (Iain Dubh Mac Iain 'Ic Ailein), who wrote popular jingles; and John Maclean (Iain Mac Ailein), who showed an interest in early Gaelic legend. Finally, bardic poetry continued to be composed into the 18th century by Niall and Domhnall MacMhuirich.

Developments of the 18th century. Almost no secular poetry in Gaelic was printed before 1751, and most earlier verse was recovered from oral tradition after that date.

The influence of Alexander Macdonald

Much of the inspiration of Gaelic printing in the 18th century can be traced to Alexander Macdonald (Alasdair Mac Mhaighstir Alasdair), who published a Gaelic vocabulary in 1741 and the first Scottish Gaelic book of secular poetry, *Ais-eiridh na Sean Chánain Albannaich* ("Resurrection of the Ancient Scottish Tongue"), in 1751. He rallied his fellow Highlanders to Prince Charles Edward's cause in the '45 rising with *Brosnachadh nam Fineachan Gaidhealach* ("Incitement to the Highland Clans") and a song of welcome to the Prince. His masterpiece, *Birlinn Chlann Raghnaill* ("The Galley of Clanranald"), is an extravaganza, ostensibly a description of a voyage from South Uist in the Hebrides Isles to Carrickfergus in County Antrim, Northern Ireland. He also composed nature poems, love poems, drinking songs, and satires.

Duncan Ban Macintyre (Donnchadh Bàn Mac an t-Saoir), who was influenced by Macdonald, had his poems published in 1768. He fought on the Hanoverian side at the Battle of Falkirk and later praised George III in *Oran do'n Rìgh* ("Song to the King"), but he had been a forester on the Perthshire–Argyllshire borders in early manhood, and this is the setting of his greatest poems, *Moladh Beinn Dóbhrainn* (*The Praise of Ben Dorain*) and *Oran Coire a Cheathaich* ("Song of the Misty Corrie"). His most famous love song is addressed to his wife, Màiri.

Other poets of note in the 18th century included John MacCodrum, author of much humorous and satirical poetry; Robert (called Rob Donn) Mackay, who wrote social satire with a wealth of shrewd and humorous understanding of human nature; and William Ross, the Romantic poet of the group, several of whose best poems, such as *Feasgar Luain* ("Monday Evening") and *Oran Eile* ("Another Song"), were occasioned by an unhappy love affair.

The greatest composer of Gaelic religious verse in the 18th century was Dugald Buchanan, who assisted the Rev. James Stewart of Killin in preparing his Gaelic translation of the New Testament (1767). His *Latha à Bhreitheanis* ("Day of Judgment") and *An Claigeann* ("The Skull") are impressive and sombre and show considerable imaginative power.

Modern trends and works. Short stories and essays appeared in 19th- and 20th-century periodicals. Alongside these were numerous religious translations from the 17th century onward, including Calvin's catechism of 1631, Gaelic translations of the Old and New Testaments, Kirk's Psalter and his Irish version of the Bible, and the 1807 Gaelic Bible. Other translations included works by John Bunyan, Richard Baxter, Thomas Boston, and Philip Doddridge. Among original prose writers were the Rev. Donald Lamont, Donald Mackechnie, and Angus Robertson. The most notable modern short stories have been written by Colin Mackenzie, John Murray, and Iain Crichton Smith.

Little vital poetry appeared in the 19th century, and a 20th-century movement to free Gaelic poetry from its traditional shackles began with Sorley Maclean, George Campbell Hay, and Derick Thomson. This has been continued by Donald MacAulay and Iain Crichton Smith.

MANX

Although they succeeded in establishing their language on the Isle of Man, the Gaels lost their hegemony over the island to the Norse in the 9th century and recovered it only from 1266 to 1333, when they lost it again to the English. They were consequently unable to provide there, as they did in Ireland and Scotland, the aristocratic support needed by the bardic institution. This, and the fact that Manx and Scottish Gaelic did not deviate significantly from Irish until the 16th century, explains why no medieval literature specifically identifiable with the island survives, and why such modern literature as exists, apart from translation literature, is predominantly folklore.

The Reformation's slow progress on the island is reflected in the comparatively late appearance of a Manx translation of the Bible and the Book of Common Prayer. The latter was completed about 1610 by a Welshman, John Phillips, bishop of Sodor and Man, but it remained unpublished until it was printed in 1893–94 side by side with the 1765 version made by the Manx clergy.

Translating the Bible into Manx was indeed a formidable task because the clergy on whom it fell had but few scholars among them and no literary tradition to draw upon. A start was made in 1748 with the appearance of a Manx version of the Gospel According to St. Matthew. A revision of Matthew and a translation of the other Gospels and of the Acts appeared in 1763, and the remainder of the New Testament in 1767. The translation of the Old Testament was published in two parts: Genesis to Esther in 1771, Job to Malachi with two books of the Apocrypha in 1773.

The Holy Scriptures were not the only religious books to be translated. Bishop Thomas Wilson's *Principles and Duties of Christianity* appeared in English and Manx in 1699, and 22 of his sermons appeared in a Manx translation in 1783. More interesting are *Pargys Caillit,* the paraphrase translation of Milton's *Paradise Lost,* which was published in 1794 and reprinted in 1872, and *Coontey ghiare yeh Ellan Vannin* ("The Short Account of the Isle of Man"), written in Manx by Joseph Bridson and printed as the 20th volume of the Publications of the Manx Language Society. As late as 1901 there appeared from the press *Skeealyn Æsop,* a selection of Aesop's fables.

More characteristic of Manx folk culture were the ballads and carols, or *carvels.* Among the most notable of the former are an Ossianic ballad describing the fate of Finn's enemy, Orree; the Manx Traditionary Ballad, a history of the island to the year 1507 made up of a mixture of fact and fiction; and the ballad on the death of Brown William; *i.e.,* William Christian, shot as a traitor in 1663. The *carvels* differ from English carols because they take as their subject not so much the Nativity as the life of Jesus, his crucifixion, and the Last Judgment. They were sung by individuals in church on Christmas Eve. With the spread of Nonconformity on the island, Manx translations of some of the popular hymns of the Methodist Revival were published.

WELSH

Taliesin and Aneirin

Literature of the Middle Ages. *Poetry.* Welsh literature has extended in an unbroken tradition from about the middle of the 6th century to the present day, but, except for two or three short pieces, all pre-Norman poetry has survived only in 12th- to 15th-century manuscripts. Welsh had developed from the older Brythonic by the middle of the 6th century. In the *Historia Brittonum* (*c.* 800) references are made to Welsh poets who, if the synchronism is correct, sang in the 6th century. Works by two of them, Taliesin and Aneirin, have survived. Taliesin wrote odes, or *awdlau,* in praise of the warlike deeds of his lord, Urien of Rheged, a kingdom in present-day southwest Scotland and northwest England. To Aneirin is attributed a long poem, *Y Gododdin,* commemorating in elegies an ill-starred expedition sent from Gododdin, the region where Edinburgh stands today, to take Catraeth (Catterick, North Yorkshire) from the invading Saxons. The background, inspiration, and social conventions of the poems of Taliesin and Aneirin are typically heroic, the language is direct and simple, and the expression terse and vigorous. These poems, and others that have not been preserved, set standards for later ages. The alliterative verse and internal rhyme found here were developed by the 13th century into the intricate system of consonant and vowel correspondence, or consonant correspondence and internal rhyme, called cynghanedd.

The heroic tradition of poetry existed also in Wales proper and was continued after the break with North Britain in the mid-7th century. The earliest surviving example is a poem in praise of Cynan Garwyn of Powys, whose son Selyf was slain in battle. This poem struck a note that remained constant in all Welsh eulogies and elegies down to the fall of the Welsh bardic system: Cynan is the bravest in the field, the most generous in his home, all others are thrall to him and sing his praises.

The period between the 7th and 10th centuries is represented by a few scattered poems, most of them in the heroic tradition, including *Moliant Cadwallon* ("The Eulogy of Cadwallon"), by Afan Ferddig, the elegy on Cynddylan ap Cyndrwyn of Powys in the first half of the 7th century, and *Edmyg Dinbych* ("The Eulogy of Tenby"), by an unknown South Wales poet. Poetry claiming to foretell the future is represented by *Armes Prydain Fawr* ("The Great Prophecy of Britain"), a stirring appeal to the Welsh to unite with other Britons, with the Irish, and with the Norse of Dublin to oppose the Saxons and to refuse the unjust demands of their "great king," probably Athelstan of Wessex. Poetry outside the main bardic tradition is preserved in englyns (stanzas of three of four lines), a dialogue between Myrddin and Taliesin, and in *Kanu y Gwynt* ("The Song of the Wind"), a riddle poem that contains the germ of the later convention known as *dyfaliad* (kenning).

The poems of Llywarch Hen

The poems associated with the name Llywarch Hen are the verse remains of at least two sagas composed toward the middle of the 9th century by an unknown storyteller of Powys, whose basic material was the traditions associated with the historical Llywarch and Heledd, sister to Cynddylan ap Cyndrwyn. In these, it seems that prose (now lost) was used for narrative and description and verse for dialogue and soliloquy. The metrical form was embellished by alliteration, internal rhyme, and incipient cynghanedd. The theme of both sagas was lamentation for the glory that once had been. The background was the heroic struggle of the Welsh of Powys against the Saxons of Mercia. Some fragments of poetry preserved in the *Black Book of Carmarthen* (12th century) were parts of soliloquies or dialogues from other lost sagas. Examples are a conversation between Arthur and the doorkeeper Glewlwyd Mightygrasp; a monologue of Ysgolan the Cleric; verses in praise of Geraint, son of Erbin; and a fragment of what may be an early native version of the Trystan and Esyllt (Tristan and Iseult) story. The manuscript shows that there once existed a legend of Myrddin Wyllt, a wild man of the woods who went mad at the sight of a battle, a legend associated with Suibne Geilt in Ireland and with Lailoken in Scotland. This Myrddin (later better known as Merlin) had the gift of prophecy. The historical poet Taliesin also became the central prophetic figure in a folk tale that was given literary form in the 9th or 10th century, but that has survived only in certain monologues preserved in *The Book of Taliesin* and in garbled versions in late texts of *Hanes Taliesin* ("Story of Taliesin").

Nature, a source of similes in the heroic poetry and of symbolism in verse fragments of the sagas, was sometimes a subject of song in its own right. Generally, treatment of the subject was remarkable for its sensitive objectivity, its awareness of form, colour, and sound, and its concise, often epigrammatic, expression. In mood, matter, and form (that of the englyn) it often overlapped with gnomic poetry, which consisted of sententious sayings about man and nature. Most gnomic and nature poems were probably produced in the 10th and 11th centuries by poets other than professional bards. Toward the end of the pre-Norman period a few poems on religious, biblical, and other subjects showed acquaintance with nonnative legends. Saga poetry gradually gave way to prose.

The poets of the princes

With the consolidation of the principality of Gwynedd under Gruffudd ap Cynan (1054–1137) and his descendants, court poetry flourished in the country, composed by the *gogynfeirdd,* or poets of the princes, who continued and developed the tradition of their predecessors, the *cynfeirdd.* The bardic order seems to have been reorganized, although no clear picture of it emerges from references in the poetry and law texts, and it seems to have been less schematized in practice than in theory. At the top of the order was the *pencerdd* ("chief of song or craft"), the ruler's chief poet, whose duty was to sing the praise of God, the ruler, and his family. Next came the *bardd teulu,* who was the poet of the ruler's war band although he seems to have been poet to the ruler's family as well. There were other, less exalted grades, with less exalted duties and the license probably to engage in satire and ribaldry.

The bardic hierarchy

Bards were also graded according to proficiency. This classification led, under Henry IV, to the holding of an eisteddfod, or a session of bards to confer certificates of proficiency and to prevent the lower orders from proliferating and drifting into mendicancy. One of the results of a bardic system of this type was a remarkable conservatism in literature. Most of the 13th-century bards used a conventional diction that was consciously archaic in its vocabulary, grammar, and idiom and incomprehensible to anyone uneducated in poetry.

Bardism often went by families, and among the first court poets were Meilyr, his son Gwalchmai, and his grandson Meilyr ap Gwalchmai, who were attached to the court of Gwynedd at Aberffraw. Gwalchmai in his *Arwyrain Owain* ("Exaltation of Owain") displayed one characteristic of all the *gogynfeirdd,* description of water, whether of river or sea. Bardic poetry, highly conventional in form, was now marked not by profundity but by adornment and linguistic virtuosity. Two poet-princes, Owain Cyfeiliog of Powys and Hywel ab Owain of Gwynedd, however, stand out from contemporary bards. Cyfeiliog's most famous work, the *Hirlas Owain* ("Owain's Long Blue Drinking Horn"), celebrates a victorious raid; Hywel ab Owain's departure from convention was more striking; for the first time in Welsh literature love of country and of beauty in the modern sense appeared: land and sea and women and the

Welsh language spoken in cultured accent by his ladylove awoke in him feelings of awe and wonder. The *gogynfeirdd* poetry alternated throughout this period between *marwnad* ("elegy") and *moliant* ("eulogy"), and the period closed with the most famous of all elegies, Gruffudd ab yr Ynad Coch's elegy after the death in 1282 of Llywelyn ap Gruffudd, last native prince of Wales.

The religious verse of the *gogynfeirdd* was generally simpler in style than the eulogies and elegies. A set type was the *marwysgafn* ("deathbed song"), in which the poet, sensing the approach of death, confessed his sins and prayed for forgiveness. Other religious poems were in praise of God and the Trinity, in honour of saints, on the torments of hell, and on the birth of Christ. They illustrate the gradual widening of the bardic horizon.

With the passing of princes and their pageantry, the poets were forced to find patrons among the new aristocracy. These patrons had more limited means and less restricted interests, with the result that the bardic system and its educational basis were gradually changed and a new kind of poetry was produced. The language became less esoteric, less specialized. Poets in the years between the English conquest (1282) and the appearance of Dafydd ap Gwilym (flourished *c.* 1340–70) seem to have returned to an earlier poetic fashion or to have been influenced by new ideas from other lands. Famous in this period of transition were Gruffudd ap Maredudd, Gruffudd ap Dafydd, and Casnodyn.

The golden age of the cywydd

The conquest of Wales by Edward I transferred the patronage of court poetry at Gwynedd and Powys from prince to landed aristocracy. The *pencerdd* lost his superiority over the lower bardic ranks, who were no longer restricted in choice of content and style, and who, especially in South Wales (where the Norman Conquest had been established for a whole century before the conquest of Gwynedd), became more vocal as the older bardic song began to decline. The new poets of the south were well established before their works began to be preserved. The most important of them was Dafydd ap Gwilym (a contemporary of Chaucer), who in his early period wrote according to two distinct traditions. He wrote *awdlau,* or odes, in the manner of the later *gogynfeirdd.* (Originally an *awdl* was a poem with a single end rhyme throughout; later it contained sequences of lines with such end rhymes. In both cases the lines were embellished with alliteration, a correspondence of consonants or internal rhyme; *i.e.,* "free" or incipient cynghanedd.) In the manner of a more popular and perhaps lower class of poets, Dafydd ap Gwilym wrote cywyddau, composed of couplets of seven-syllable lines rhyming in alternately stressed and unstressed final syllables. Each line was embellished by a stricter form of cynghanedd, the *cynghanedd gaeth.* Dafydd established, if he did not invent, the cywydd, but his main achievement was the simplicity of diction he cultivated in it. His successors followed his lead; the old diction became obsolete, and he thus established the standards of modern Welsh. The substance of his poetry was also new, for he seems to have borrowed many of his themes from the wandering minstrels and trouvères of France. He wrote love poetry but perhaps is best known for his descriptions of nature.

Dafydd ap Gwilym's influence

Dafydd's influence was twofold: the cywydd was established as the leading form, and the new subjects were recognized as fit themes for poetry. One contemporary, Gruffudd ab Adda, went much further toward a modern conception of nature; another, Iolo Goch, in his poem to the husbandman shows traces of English ideas, as seen in *Piers Plowman.* Llywelyn Goch Amheurug Hen wrote some early poems in the *gogynfeirdd* tradition, but his "Elegy to Lleucu Llwyd" successfully combined the Welsh elegy tradition with the imported serenade form.

In the 15th century the cywydd was refined. Although Dafydd Nanmor was inferior to most of Dafydd ap Gwilym's contemporaries in treatment of his subject and in imagination, in his mastery of the cywydd form he had no equal. Further advances in the cywydd metre were made by Lewis Glyn Cothi and Guto'r Glyn, in whose work a real consciousness of Welsh nationhood is seen.

Prose. The earliest examples of Welsh prose were utilitarian: notes on Latin texts dealing with weights and measures, an agreement, a list of church dues, and an astronomical commentary. Shortly before the middle of the 10th century Howel (Hywel) Dda, according to tradition, had the Welsh laws codified. Although the earliest Welsh law manuscripts belong to the first half of the 13th century, some of the texts probably derive from 12th-century and earlier exemplars.

The *Mabinogion* collection of tales

The stylistic merits of the legal texts were reflected in a more conscious literary use of prose by storytellers (*cyfarwyddiaid*), who recited oral tales made up of a medley of mythology, folklore, and heroic elements. Some of these were recorded in writing; the most famous collection is the *Mabinogion,* preserved in *The White Book of Rhydderch* (*c.* 1300–25) and *The Red Book of Hergest* (*c.* 1375–1425). The *Mabinogion* are composed of 11 anonymous tales, based on older oral material. The greatest are the four related stories "The Four Branches of the *Mabinogi,*" redacted in the second half of the 11th century by a writer from Dyfed. The author of "Culhwch and Olwen" (*c.* 1100), drawing on material akin to that of the "Four Branches," appears to have kept closer to the oral tale, but his inferior stylistics presaged the later decadent *areithiau* ("rhetorics"), which were in part parodies of the *Mabinogion.* Three of the *Mabinogion* tales, "Owain and Luned" (or "The Lady of the Fountain"), "Geraint and Enid," and "Peredur Son of Efrawg," represented a transition from purely native tales to those composed under Norman influence. These romances correspond to the *Yvain, Erec,* and *Perceval* of Chrêtien de Troyes, and the exact relationship between the Welsh and French texts has long been debated. Although more sophisticated, they show a decline from the directness and restraint of the native tales.

Many translations from Latin and French helped to create a prose that could express aspects of life rarely touched on in the tales. Most of them dated from the 13th and 14th centuries and were probably made by monks. Notable among them were translations from the Latin of Geoffrey of Monmouth's *Historia regum Britanniae* and the French of the *Queste del Saint Graal* and *Perlesvaux.* Their prose was largely experimental and influenced in varying degrees by the language and style of the originals, but at its best it was a not unworthy development of the prose of the law tracts and native tales.

The development of medieval prose was hampered by a gap between Welsh literary tradition and the wider learning of western Europe. The inspiration for the fashioning of a prose able to express all facets of thought and activity arose toward the middle of the 16th century out of the Reformation and the Renaissance, from foundations laid by humanists, both Protestant and Catholic.

The Reformation and the Renaissance. After the outstanding period of Dafydd ap Gwilym and his followers in the cywydd, there arose for a short time a school of literary formalists. The chief of these was Dafydd ab Edmwnd, whose poetic heirs were Tudur Aled (died 1526) and Gutun Owain (flourished *c.* 1460–1500).

Effects of the Reformation and the accession of the Tudors

The Reformation broke the hold of the Roman Catholic religion on Welsh life without establishing at the same time a similar hold of its own. The Tudor policy of encouraging the spread of English at the expense of Welsh and of inducing the Welsh aristocracy to emigrate to England almost destroyed the old Welsh culture, completely bound up as it was with the language. Yet fine poetry was written by the satirists Siôn Tudur and Edmwnd Prys. Other masters of the cywydd were William Llŷn and Siôn Phylip.

The rise of modern prose. When printing began in Wales in the 16th century, traditional prose was abandoned by the Renaissance humanists. The new prose they fashioned was based on bardic language and classical authors, enriched by new formations and borrowings. The first Welsh printed book, *Yn y lhyvyr hwnn* (1547; "In This Book"), consisted of extracts from the Scriptures and the prayer book: from this time modern Welsh prose began to assume definite form.

The Reformation. The most important figure of the Reformation was William Salesbury, who translated most of the New Testament of 1567. Despite some eccentricities, it was a fine piece of translation. In the same year was published the Welsh Prayer Book, also translated mainly by Salesbury in collaboration with Richard Davies, bishop of St. David's. The Welsh Bible translated by William Morgan, bishop of St. Asaph, aided by Edmwnd Prys, was published in 1588. The revised version, published in 1620, is still used. It would be difficult to exaggerate the importance of these three translations in the development of Welsh prose. They started a steady, if not large, stream of Welsh prose books. The first were translations from English and Latin aimed at grounding the Welsh nation in the principles of the Reformation.

The Counter-Reformation. While the reformed religion was being established in Wales, Welsh society and the Welsh language were at their lowest ebb. The Roman Catholic writers of the Counter-Reformation regarded the new religion as an English import and struggled to preserve old Roman Catholic culture. As a result there appeared *Dosparth Byrr* ("A Short Rationale"), the earliest printed Welsh primer, the work of Gruffydd Robert (*c.* 1522–*c.* 1610), and several religious works, many of which were published on the Continent.

The Welsh Renaissance. Just as Italy and other European countries during the Renaissance turned to the Latin and Greek classics, so Wales turned to its own classical tradition of bardism. In addition to Gruffydd Robert's primer mentioned above, there appeared a set of rules for bardic poetry and principles of the Welsh language compiled by Siôn Dafydd Rhys and a dictionary and a grammar by John Davies of Mallwyd.

Religious emphasis in Welsh writing

Welsh literature in the 17th century. *Prose.* So far, writers of Welsh prose had contented themselves with translation, until Morgan Llwyd produced his religious works. A Puritan, he made an original contribution to Welsh religious thought, chiefly in *Llyfr y Tri Aderyn* (1653; "The Book of the Three Birds"), a disquisition on government and religious liberty, and *Llythur ir Cymru Cariadus* (*c.* 1653; "Letter to the Beloved Welsh"), which expounded a mystical gospel. Among the clergy who produced some of the many translations, mostly of religious originals, during this period were Edward Samuel; Moses Williams, a diligent searcher into manuscripts; Griffith Jones, the father of Welsh popular education; and Theophilus Evans, author of *Drych y Prif Oesoedd* (1716; *A View of the Primitive Ages*). Ellis Wynne o Lasynys is often regarded as the greatest of Welsh prose writers. His two great works were *Rheol Buchedd Sanctaidd* (1701), a translation of Jeremy Taylor's *Rule and Exercises of Holy Living,* and *Gweledigaetheu y Bardd Cwsc* (1703; *The Visions of the Sleeping Bard*), an adaptation of a translation of the *Sueños* of the Spanish poet Quevedo.

Poetry. When Henry VII came to the throne, the old Welsh gentry turned toward England for preferment. Soon, poets of the older school had no audience, and only the rich gentlemen farmers kept up the old tradition. A new school, however, was rising that combined a vast store of folk song, previously despised and unrecorded, with imitation of contemporary English popular poetry and sophisticated lyrics. Landmarks of this new development were Edmwnd Prys's metrical version of the Psalms and Rhys Prichard's *Canwyll y Cymry* (1646–72; "The Welshman's Candle"), both written in so-called free metres. Prys's Psalter contained the first Welsh metrical hymns. Prichard's work consisted of moral verses in the metres of the old folk songs (*penillion telyn*). Many other poets wrote in these metres, but they were generally crude until handled by the greatest poet of the period, Huw Morus, who was particularly famous for his love poems. Later came Lewis Morris, the inspirer and patron of Goronwy Owen and thus a strong link with the next extremely productive period.

Owen's poetic reforms

The 18th century: the first revival. The mid-18th century was, after the 14th, the most fruitful period of Welsh literature. Goronwy Owen, inspired by English Augustanism, reintroduced and improved the strict metres of the cywydd and *awdl* (by this time a long poem written in a number of the classical cynghanedd metres). He also introduced a wide range of subject content, and thus founded a new classical school of Welsh poetry. The more important poets of this school were William Wynn of Llangynhafal, Edward Richard, and Evan Evans (Ieuan Fardd). Much of their activity was associated with the Welsh community in London and the Cymmrodorion Society and led to the establishment of local eisteddfods in Wales, which perpetuated the classical forms of Welsh poetry.

Chief among Owen's successors was David Thomas (Dafydd Ddu Eryri), who, however, like other eisteddfodic bards of this period, soon departed from classical strictness.

The classicists of the 18th century stood aloof from the Methodist Revival, but religious fervour brought a new articulateness and inspired poets in free metre, especially hymn writers like the preeminent William Williams of Pant-y-celyn and the mystical Ann Griffiths.

For a long time after 1750, Welsh prose was mainly concerned with religious subjects. The French Revolution, however, gave impetus to political writing, and among those it influenced was John Jones (Jac Glan-y-gors). It was only after a periodical press had been established that politics began to compete with religion as a subject for comment.

19th-century literary trends. *Poetry.* Strict metre poetry declined in the 19th century. Although the volume produced was enormous, the quality was poor. Eben Fardd was probably the last of the eisteddfodic poets to make any real contribution. The influence of the hymn writers of the 18th century was seen in the development of the lyric. In fact, all the poetry of the 19th century betrays a religious origin. The influence of contemporary English songs was also seen, as in the work of John Blackwell (Alun). More originality was shown by Evan Evans (Ieuan Glan Geirionydd), who founded the Eryri school of poetry, inspired by the scenery of Snowdonia. The earlier lyricists were followed by a more bohemian group—Talhaiarn (John Jones), Mynyddog (Richard Davies), and Ceiriog (John Hughes), the latter the greatest lyrical writer of the century. Only one poet, Islwyn (William Thomas), made a success of the long poem: his *Y Storm* is a series of meditations on life and art.

Prose. This was the most prolific period of Welsh prose, though much of it was of poor quality, partly because it was produced by a people who had had little formal education in their own language and who had lost touch with their own literary past. Much of it was marred also by the pretentious style of the followers of William Owen Pughe, who tried to "restore" literary standards. A tremendous volume of prose was produced—periodicals, religious books and tracts, biographies, sermons, letters, and monumental works such as *Y Gwyddoniadur* ("The Encyclopaedia") and *Hanes y Brytaniaid a'r Cymry* ("History of the Britons and the Welsh"). Political writings became an important part of Welsh literature, the two great political writers of the century being Samuel Roberts and Gwilym Hiraethog (William Rees). Lewis Edwards, founder and editor of *Y Traethodydd* ("The Essayist"), tried to introduce a wider, European standard of literary criticism. There were some interesting attempts at creative writing, but the only great novelist was Daniel Owen, whose work portrays the extraordinary influence of religion on contemporary society.

Welsh literary renaissance

The second revival. The most important event for the second revival in Welsh literature was the establishment of the University of Wales (1872–93). The immediate result was a great widening of literary horizons, accompanied by a strong reaction toward the old Welsh classical ideas. Sir Owen M. Edwards, Sir John Morris-Jones, Emrys ap Iwan (Robert Ambrose Jones), and others made the Welsh conscious of their literary identity and set new standards for correctness of language and integrity of thought. The great literary renaissance that followed was marked by T. Gwynn Jones's masterly use of the old strict metres to express modern thought and W.J. Gruffydd's lyrical use of the free metres to express his rebellion against society and his love for the countryside of his youth. R. Williams Parry showed a superb gift of poetic observation, while Sir Thomas Parry-Williams combined a mystical love for his native Gwynedd with an almost scientific analysis of his own metaphysical preoccupations. Older poets, such as Cynan (A. Evans-Jones), William Morris, and Wil Ifan (William Evans), clung to earlier lyrical models, although many others, like D. Gwenallt Jones and Saunders Lewis, drew increasingly on the rhythms and vocabulary of colloquial speech. Waldo Williams, Gwilym R. Jones, the younger Bobi Jones, and particularly Euros Bowen experimented with form and subject. Their work has been followed and developed by writers of the younger generation, the most distinctive and prolific being Gwyn Thomas. Interest in the use of the strict metres of cynghanedd has revived, as reflected in the publication of the popular periodical *Barddas* ("Bardism"). Most poets, old and young, reflect a varying involvement in contemporary Welsh political activity.

The high standard of the periodical *Y Llenor* ("The Litterateur"; 1922–51) indicated the advances made in prose. Contributors were generally involved in a wide range of activities: its editor, W.J. Gruffydd, was both poet and essayist; Saunders Lewis was a poet, dramatist, and politician; Sir Thomas Parry-Williams a poet and essayist; and R.T. Jenkins an essayist and historian. Together with novelists and short-story writers like Tegla Davies, T. Rowland Hughes, Kate Roberts, and D.J. Williams, they effectively mirrored contemporary Wales. John Gwilym Jones and Islwyn Ffowc Ellis were innovators in form and subject matter, and they were followed by an enthusiastic younger generation. Literary criticism has also benefited. The standard set by *Y Llenor* has been maintained in *Ysgrifau Beirniadol* ("Critical Essays"). In this field as in others the establishment of the Welsh Academy (Yr Academi Gymreig) in 1959 and the publication of its review *Taliesin* have made an outstanding contribution.

Drama in Wales was first written in the 20th century. At first realistic, it developed into poetic, symbolic drama, often based on historical and mythological themes but dealing with moral, social, and psychological contemporary problems. The outstanding names are Saunders Lewis, John Gwilym Jones, Emyr Humphreys, and Gwenlyn Parry.

CORNISH

The oldest remains of Cornish are proper names in the Bodmin Gospels and in the Domesday Book, 10th-century glosses on Latin texts, and a 12th-century vocabulary based on Aelfric's Latin–Anglo-Saxon glossary. The earliest literary text in Middle Cornish is a 41-line fragment of a drama, written about 1400, in which a girl is offered as wife, praised for her virtues, and counseled on her behaviour. The other plays that have survived complete or in part are related to the medieval miracle and morality plays.

Of these the most important is the *Ordinalia,* a trilogy written in Middle Cornish (probably late 14th century) and designed to be acted on three consecutive days, perhaps in a *plen-an-gwary* ("play-field"). The first play, the *Origo Mundi* ("Origin of the World"), is based on the Old Testament and serves as a prologue to the other two, the *Passio Domini Nostri Ihesu Christi;* and the *Resurrexio Domini Nostri Ihesu Christi.* Lacking a treatment of the Nativity and of the ministry of Jesus, the scheme underlying the *Ordinalia* is more like that of the great French Passions than that of the English Corpus Christi cycle.

The *Pascon Agan Arluth* ("Passion of Our Lord"), of which 259 stanzas have survived, has passages common to the *Ordinalia* and must have been composed about the same time.

Beunans Meriasek ("The Life of Meriasek [or Meriadoc]"), on the patron saint of Camborne, is held to be considerably later: the manuscript in which it is found was completed in 1504 by one Dominus Hadton, thought to have been a canon of the Collegiate Church of Glasney. *Gwreans an Bys* ("The Creation of the World"), surviving in a text written in 1611, was intended to be the first of three plays and seems to be an extended and reworked version of the *Origo Mundi.* Written in Late Cornish, it is the latest of the Cornish religious plays.

The Protestant Reformation destroyed the mentality that had created and enjoyed these plays, although there is evidence that the destruction was slow. Unfortunately the Reformation provided no work in the Cornish language that could take the place of the plays. A few people recognized that the Bible and the Book of Common Prayer should be made accessible to the monoglot Cornishmen, but little was done to that end. John Tregear translated into Cornish 12 of Bishop Edmund Bonner's sermons and a 13th sermon from an unidentified source, and it is probably significant that this translation was lost and not rediscovered until 1949.

Attempts have been made to revive the use of Cornish as a spoken and written language. Among the societies and movements that have worked to this end are the Cornish-Celtic Society, The Old Cornwall Society, Land and Language, The Sons of Cornwall, and The Cornish Language Board; and among the papers or periodicals these have promoted, *Kernow* ("Cornwall"), *An Lef* ("The Voice"), *An Lef Kernewek* ("The Cornish Voice"), and *Hedhyu* ("Today") have been notable for the attention they have given to compositions in Cornish—mostly poems, short stories, and short plays. Among the Cornish language revivalists are Henry Jenner, R. Morton Nance, and A.S.D. Smith.

BRETON

The three major periods of Breton literature. Old Breton (8th to 11th century) is found only in lists and glosses in documents or as names in Latin books and charters. From the Middle Breton period (11th to 17th century) the 11th- to 15th-century compositions were mainly oral, and little except a few scraps of verse is extant until the late 15th century, when there appeared the *Catholicon* of Jean Lagadeuc, a Breton–Latin–French dictionary printed in 1499, and Quiquer de Roscoff's French–Breton dictionary and conversations (printed 1616).

A 17th-century collection, *Cantiques bretons* (1642), names several Breton airs. All the remaining works of the middle period were religious and mostly in verse. Three mystery plays were probably the most significant products of the period: *Buez santez Nonn* ("Life of St. Nonn"), *Burzud bras Jesuz* ("The Great Mystery of Jesus"), and *Buhez santes Barba* ("Life of St. Barbara"). Three long poems, "Tremenvan an itron gwerches Maria" ("The Passion of the Virgin Mary"), "Pemzec levenez Maria" ("The Fifteen Joys of Mary"), and "Buhez mabden" ("The Life of Man"), were all probably based on French versions. A book of hours in verse, a prose extract from the Leon missal, and a prose catechism belong to this period, as does the prose *Buhez an itron sanctes Cathell, guerches ha merzeres* ("Life of St. Catherine, Virgin and Martyr"). *Am Mirouer a Confession* ("The Mirror of Confession") and *Doctrin an Christenien* ("Christian's Doctrine") are translated from the French. A collection of carols, *An Nouelou ancient ha devot* ("Ancient and Devout Songs"), appeared in 1650, and a book of metrical meditations in 1651. In general, Middle Breton literature lacked originality, and the indigenous culture of Brittany seems to have been entirely neglected by the educated classes, who introduced an enormous number of French words into the vocabulary.

Modern Breton is said to have begun in 1659, when Julien Maunoir introduced a more phonetic orthography, but works of the Middle Breton type appeared until the 19th century. The bulk of Breton literature in this period consisted of mystery and miracle plays treating subjects from the Old and New Testaments, saints' lives, and stories of chivalry derived from French or Latin. Even plays depicting Breton saints evinced little originality. In the 18th century many Breton dictionaries were published but little of literary significance was produced. One name survives: Claude-Marie Le Laë, who wrote satiric poems.

The revival of Breton literature. Interest in Breton, which revived at a time when France's central government was trying to impose French on Brittany and destroy the regional language, was particularly stimulated with the publication of the celebrated *Barzaz Breiz* (originally *Barzas-Breiz,* 1839; "Breton Bardic Poems"). This collection of poems was compiled by Théodore Hersart de La Villemarqué, who declared that they had survived unchanged as part of Breton folklore. Breton-speaking scholars doubted the collection's authenticity, and attacks reached their height when R.-F. Le Men, in a reprinting

Barzaz Breiz

in about 1870 of *Catholicon,* and François-Marie Luzel, in a paper delivered in 1872, showed that *Barzaz Breiz* was not authentic (though scholars during the period often edited such collected material). *Barzaz Breiz* led to a renaissance of Breton writing and stimulated Luzel to collect authentic folk songs and publish *Gwerziou Breiz-Izel* (2 vol., 1868–74; "Ballads of Lower Brittany") and, in collaboration with Anatole Le Braz, *Soniou Breiz-Izel* (2 vol., 1890; "Folk Songs of Lower Brittany"). In the 1980s Donatien Laurent, the first to have had access to Villemarqué's papers, demonstrated that some of the poems were authentic.

Prose. Luzel also collected folktales and legends, publishing many in Breton as well as in French translation. His collaborator, Le Braz, published stories concerning *an ankou* ("death"), as *La Légende de la mort* (1893; *Dealings with the Dead*). Traditional and literary elements combined indistinguishably in many stories. When Breton writers did not depend on folk legends for material, they fictionalized their own life stories. The many improving religious works published were not at all original; yet many Bretons who have read only one book in their own language have read *Buez ar Zent* ("Lives of the Saints").

Most playwrights were concerned to teach moral and religious lessons, such as Toussaint Le Garrec and Abbé J. Le Bayon, who revived several great mystery plays—*Nicolazig, Boeh er goed* ("The Voice of the Blood"), *Ar hent en Hadour* ("In the Steps of the Sower"), and *Ar en hent de Vethleem* ("On the Way to Bethlehem").

Poetry. For 200 years Bretons expressed their feelings in poems that were published as pamphlets—either as *soniou* (love songs, satires, carols, marriage lays) or *gwerziou* (ballads or broadsides describing recent events in Brittany and elsewhere). Their authors were people of every social class, and the poems were hawked at fairs. Dozens of poets published collections after the appearance of *Barzaz Breiz.*

20th-century Breton writers

Developments of the 20th century. Collections of folklore continued to appear, but there was an ever-increasing realization that the reading public needed a literature reflecting contemporary life. Literary periodicals attempted to satisfy the need. Worthy of special mention are *Dihunamb* (1905; "Let Us Wake Up"), *Gwalarn* (1925–44; "North West"), and *Al Liamm* (1946; "The Bond"). Tangi Malemanche was a prolific playwright but gained little recognition. Per-Jakez Helias as poet, playwright, and radio script writer has been both prolific and popular. The reminiscences of Yeun ar Go in *E Skeud Tour Bras Sant Jermen* (1955; "In the Shadow of the Great Tower of Saint-Germain") and continued in *Al Liamm* have great charm. But the three writers who can claim to have done most to give Brittany a 20th-century Breton literature have been Jakez Riou and Youenn Drezen, authors of short stories and novellas, and Roparz Hemon, who has enriched every genre. Maodez Glanndour has given of his best to poetry. As one would expect in a literature in which little prose was written until the 20th century, verse continues to attract the attention and the skill of young and old writers. One of its recurring themes is the fear that the Breton language will die.

(D.Gr./D.S.T./T.Jo./J.E.C.W.)

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(J.E.C.W.)

Censorship

To censor is to act so as to change or suppress speech or writing that is condemned as subversive of the common good. This article discusses censorship and its relation to government and the rule of law.

The article is divided into the following sections:

Concerns relevant to censorship

THE STATUS OF "INDIVIDUALITY"

Censorship, as a term in English, goes back to the office of censor established in Rome in 443 BC. That officer, who conducted the census, regulated the morals of the citizens counted and classified. But, however honourable the origins of its name, censorship itself is today generally regarded as a relic of an unenlightened and much more oppressive age.

Illustrative of this change in opinion is how a community responds to such a sentiment as that with which Protagoras (*c.* 485–410 BC) opened his work *Concerning the Gods:*

> About the gods I am not able to know either that they are, or that they are not, or what they are like in shape, the things preventing knowledge being many, such as the obscurity of the subject and that the life of man is short.

This public admission of agnosticism scandalized Protagoras' fellow Greeks. Such statements would no doubt have been received with hostility, and probably with social if not even criminal sanctions, throughout the ancient world. There are few places in the modern world, on the other hand, where such a statement cannot be made without the prospect of having to endure a pained and painful community response. This change reflects, among other things, a profound shift in opinion as to what is and is not a legitimate concern of government.

Whereas it could once be maintained that the law forbids whatever it does not permit, it is now generally accepted—at least wherever Western liberalism is in the ascendancy—that one may do whatever is not forbidden by law. Furthermore, it is now believed that what may be properly forbidden by law is quite limited. Much is made of permitting people to do with their lives (including their opinions) as they please, so long as they do no immediate and evident (usually physical) harm to others. Thus, Leo Strauss has observed, "The quarrel between the ancients and the moderns concerns eventually, and perhaps even from the beginning, the status of 'individuality.' "

All this is to say that individualism is made much of in modernity. The status, then, of censorship very much depends on the standing of government itself and of legitimate authority, revealing still another aspect of the complicated relation between "the individual and the state."

REQUIREMENTS OF SELF-GOVERNMENT

One critical source of the contemporary repudiation of censorship in the West depends on something that may be distinctive to modernity, an emphasis upon the dignity of the individual. This respect for individuality has its roots both in Christian doctrines and in the (not unrelated) sovereignty of the self reflected in state-of-nature theories about the foundations of social organization. Vital to this approach is the general opinion about the nature and sanctity of the human soul. This general opinion provides the foundation of a predominantly new, or modern, argument against censorship—against anything, in fact, that interferes with self-development, and especially such self-development (or, better still, "self-fulfillment") as a person happens to want and to choose for himself. This can be put in terms of liberty—the liberty to become and to do what one pleases.

The old, or traditional, argument against censorship was much less individualistic and much more political in its orientation, making more of another sense of liberty. According to that sense, if a people is to be self-governing, it must have access to all information and arguments that may be relevant to its ability to discuss public affairs fully and to assess in a competent manner the conduct of the officials it chooses. Thus, "freedom of speech," which is constitutionally guaranteed to the people of the United States, first comes to view in Anglo-American legal history as a guarantee for the members of the British Parliament assembled to discuss the affairs of the kingdom.

Origin of freedom of speech

In the circumstances of a people actually governing itself, it is obvious that there is no substitute for freedom of speech and of the press, particularly as that freedom permits an informed access to information and opinions about political matters. Even the more repressive regimes today recognize this underlying principle, in that their ruling bodies try to make certain that they themselves become and remain informed about what is "really" going on in their countries and abroad, however repressive they may be in not permitting their own people to learn about and openly to discuss public affairs. Whether anyone who thus rules unjustly, or otherwise improperly, can be regarded as truly understanding and hence truly controlling his situation is a question not limited to these circumstances.

"FREEDOM OF EXPRESSION"

The shift from the more political to the more individualistic view of liberty may be seen in how the constitutional guarantees with respect to speech and the press are typically spoken of in the United States today. Restraints upon speaking and publishing, and indeed upon action generally, are fewer than at any time in the history of the country. This absence of restraints is reflected as well in the very terms in which these rights and privileges are described. What would once have been referred to as "freedom of speech and of the press" (drawing upon the language of the First Amendment to the Constitution of the United States) is now often referred to as "freedom of expression."

To make much of freedom of expression is to encourage a liberation of the self from the constraints of the community. It may even be to assume that the self has, intrinsic to it or somehow available to it independent of any social guidance, intimations of what it is and what it wants. Thus, liberation may be seen in the currently fashionable insistence upon "doing one's own thing"—which can include a reliance upon standards and objectives that are solely one's own. It is difficult, in such circumstances, to avoid a radical subjectivism that tends to result in a thoroughgoing relativism with respect to moral and political judgments. One consequence of this approach is to identify an ever-expanding array of forms of expression that are believed to be entitled to immunity from government regulation.

On the other hand, if the emphasis is placed upon the more traditional language, "freedom of speech and of the press," the requirements and prerogatives of a self-governing people are apt to be made more of. This means, among other things, that a people must be prepared and equipped to make effective use of its considerable political power. (Even those rulers who act without the authority of the people must take care to shape their people in accordance with the needs and circumstances of their regime. This kind of effort need not be altogether selfish on the part of such rulers, since all regimes do have an interest in law and order, in common decency, and in a routine reliability or loyalty.) It should be evident that a people entrusted with the power of self-government must be able to exercise a disciplined judgment: not everything goes, and there are better and worse things awaiting the community and its citizens.

True freedom

What is particularly difficult to argue for, and to maintain, is an arrangement that, while it leaves a people clearly free politically to discuss fully all matters of public interest with a view toward governing itself, routinely disciplines and otherwise prepares that same people for an effective exercise of its considerable freedom. In such circumstances, it is tempting for a people to take the case for, and the rhetoric of, liberty one step further—and thus to insist that no one should try to tell anyone else what kind of human being he should be. Yet, cannot it be said that only they are truly free who know what they are doing and who choose to do what is right? All others are, in varying degrees, prisoners of illusions and appetites, however much they may believe themselves to be freely expressing themselves.

There are, then, two related sets of concerns evident in any consideration today of the forms and uses of censorship. One set of concerns has to do with the everyday governance of the community; the other, with the permanent shaping of the character of the people. The former is more political in its methods, the latter is more educational.

History of censorship

It should be instructive to consider how the problem of censorship has been dealt with in the ancient world, in premodern times, and in the modern world. Care must be taken here not to assume that the modern democratic regime, of a self-governing people, is the only legitimate regime. Rather, it is prudent to assume that most of those who have, in other times and places, thought about and acted upon such matters have been at least as humane and as sensible in their circumstances as modern democrats are apt to be in theirs.

ANCIENT GREECE AND ROME

It was taken for granted in the Greek communities of antiquity, as well as in Rome, that citizens would be formed in accordance with the character and needs of the regime. This did not preclude the emergence of strong-minded men and women, as may be seen in the stories of Homer, of Plutarch, of Tacitus, and of the Greek playwrights. But it was evident, for example, that a citizen of Sparta was much more apt to be tough and unreflective (and certainly uncommunicative) than a citizen of Corinth (with its notorious openness to pleasure and luxury).

Power of religious opinion

The scope of a city's concern was exhibited in the provisions it made for the establishment and promotion of religious worship. That "the gods of the city" were to be respected by every citizen was usually taken for granted. Presiding over religious observances was generally regarded as a privilege of citizenship: thus, in some cities it was an office in which the elderly in good standing could be expected to serve. A refusal to conform, at least outwardly, to the recognized worship of the community subjected one to hardships. And there could be difficulties, backed up by legal sanctions, for those who spoke improperly about such matters. The force of religious opinions could be seen not only in prosecutions for refusals to acknowledge the gods of the city but perhaps even more in the frequent unwillingness of a city (no matter what its obvious political or military interests) to conduct public business at a time when the religious calendar, auspices, or other such signs forbade civic activities. Indicative of respect for the proprieties was the secrecy with which the religious mysteries, such as those into which many Greek and Roman men were initiated, were evidently practiced—so much so that there does not seem to be any record from antiquity of precisely what constituted the various mysteries. Respect for the proprieties may be seen as well in the outrage provoked in Sparta by a poem by Archilochus (7th century BC) in which he celebrated his lifesaving cowardice.

Athens, it can be said, was much more liberal than the typical Greek city. This is not to suggest that the rulers of the other cities did not, among themselves, freely discuss the public business. But in Athens the rulers included much more of the population than in most cities of antiquity—and freedom of speech (for political purposes) spilled over there into the private lives of citizens. This may be seen, perhaps best of all, in the famous funeral address given by Pericles in 431 BC. Athenians, he pointed out, did not consider public discussion merely something to be put up with; rather, they believed that the best interests of the city could not be served without a full discussion of the issues before the assembly. There may be seen in the plays of an Aristophanes the kind of uninhibited discussions of politics that the Athenians were evidently accustomed to, discussions that could (in the license accorded to comedy) be couched in licentious terms not permitted in everyday discourse.

The limits of Athenian openness may be seen, of course, in the trial, conviction, and execution of Socrates in 399 BC on charges that he corrupted the youth and that he did not acknowledge the gods that the city did but other new divinities of his own. One may see, as well, in the *Republic* of Plato an account of a system of censorship, particularly of the arts, that is comprehensive. Not only are various opinions (particularly misconceptions about the gods and about the supposed terrors of death) to be discouraged, but various salutary opinions are to be encouraged and protected without having to be demonstrated to be true. Much of what is said in the *Republic* and elsewhere reflects the belief that the vital opinions of the community could be shaped by law and that men could be penalized for saying things that offended public sensibilities, undermined common morality, or subverted the institutions of the community.

The circumstances justifying the system of comprehensive "thought control" described in Plato's *Republic* are obviously rarely to be found. Thus, Socrates himself is recorded in the same dialogue (and in Plato's *Apology*) as recognizing that cities with bad regimes do not permit their misconduct to be questioned and corrected. Such regimes should be compared with those in the age of the good Roman emperors, the period from Nerva (*c.* AD 30–98) to Marcus Aurelius (121–180), the golden times, said Tacitus, when everyone could hold and defend whatever opinions he wished.

ANCIENT ISRAEL AND EARLY CHRISTIANITY

Much of what can be said about ancient Greece and Rome could be applied, with appropriate adaptations, to ancient Israel. The stories of the difficulties encountered by Jesus, and the offenses he came to be accused of, indicate the kinds of restrictions to which the Jews were subjected with respect to religious observances and with respect to what could and could not be said about divine matters. (The inhibitions so established were later reflected in the manner in which Moses Maimonides [1135–1204] proceeded in his publications, often relying upon "hints" rather than upon explicit discussion of sensitive topics.) The prevailing watchfulness, lest someone say or do what he should not, can be said to be anticipated by the commandment "You shall not take the name of the Lord your God in vain; for the Lord will not hold him guiltless who takes his name in vain" (Ex. 20:7). It may be seen as well in the ancient opinion that there is a name for God that must not be uttered.

It should be evident that this way of life—directing both opinions and actions and extending down to minute daily routines—could not help but shape a people for centuries, if not for millennia, to come. But it should also be evident that those in the position to know, and with a duty

Nathan the Prophet

to act, were expected to speak out and were, in effect, licensed to do so, however cautiously they were obliged to proceed on occasion. Thus, the prophet Nathan dared to challenge King David himself for what he had done to secure Bathsheba for his wife (II Sam. 12:1–24). On an earlier, perhaps even more striking, occasion, the patriarch Abraham dared to question God about the terms on which Sodom and Gomorrah might be saved from destruction (Gen. 18:16–33). God made concessions to Abraham, and David crumbled before Nathan's authority. But such presumptuousness on the part of mere mortals is possible, and likely to bear fruit, only in communities that have been trained to share and to respect certain moral principles grounded in thoughtfulness.

The thoughtfulness to which the Old Testament aspires is suggested by the following counsel by Moses to the people of Israel (Deut. 4:5–6):

> Behold, I have taught you statutes and ordinances, as the Lord my God commanded me, that you should do them in the land which you are entering to take possession of it. Keep them and do them; for that will be your wisdom and your understanding in the sight of the peoples, who, when they hear all these statutes, will say, "Surely this great nation is a wise and understanding people."

This approach can be considered to provide the foundation for the assurance that has been so critical to modern arguments against censorship (John 8:32): ". . . and you will know the truth, and the truth will make you free." Further biblical authority against censorship may be found in such "free speech" dramas as that described in Acts 4:13–21.

It should be remembered that to say everything one thought or believed was regarded by pre-Christian writers as occasionally irresponsible or licentious: social consequences dictated a need for restraint. Christian writers, however, called for just such saying of everything as the indispensable witness of faith: transitory social considerations were not to impede, to the extent that they formerly had, the exercise of such a liberty, indeed of such a duty, so intimately related to the eternal welfare of the soul. Thus we see an encouragement of the private—of an individuality that turned eventually against organized religion itself and legitimated a radical self-indulgence.

ANCIENT CHINA

Perhaps no people has ever been so thoroughly trained, on such a large scale and for so long, as the Chinese. Critical to that training was a system of education that culminated in a rigorous selection, by examination, of candidates for administrative posts. Particularly influential was the thought of Confucius (551–479 BC), with its considerable emphasis upon deference to authority and to family elders and upon respect for ritual observances and propriety. Cautiousness in speech was encouraged, licentious expressions were discouraged, and long-established teachings were relied upon for shaping character. All in all, it was contrary to Chinese good taste to speak openly of the faults of one's government or of one's rulers. And so it could be counselled by Confucius, "He who is not in any particular office has nothing to do with plans for the administration of its duties" (*Analects* [*Lun yü*], 7:14). It has been suggested that such sentiments have operated to prevent the spread in China of opinions supportive of political liberty.

Confucius

Still, it could be recognized by Confucius that "oppressive government is fiercer than a tiger." He could counsel that if a ruler's words are not good, and if people are discouraged from opposing them, the ruin of the country can be expected (*Analects,* 13:5). Blatant oppressiveness, and an attempt to stamp out the influence of Confucius and of other sages, could be seen in the wholesale destruction of books in China in 231 BC. But the Confucian mode was revived thereafter, to become the dominant influence for almost two millennia. Its pervasiveness may well be judged oppressive by contemporary Western standards, since so much depended, it seems, on mastering the orthodox texts and discipline.

Whether or not the typical Chinese government was indeed oppressive, effective control of information was lodged in the authorities, since access to the evidently vital public archives of earlier administrations was limited to a relative few. In addition, decisive control of what was thought, and how, depended in large part on a determination of what the authoritative texts were—something that has been critical in the West, as well, in the establishment of useful canons, both sacred and secular. Thus, Richard McKeon has suggested, "Censorship may be the enforcement of judgments based on power, passion, corruption, or prejudice—political, popular, elite, or sectarian. It may also be based on scholarship and the use of critical methods in the interest of advancing a taste for literature, art, learning, and science."

MEDIEVAL CHRISTENDOM

Among the heirs of Greece and Rome and of Israel were the Christians of varying professions. Perhaps the most dramatic form of censorship in Christendom was that displayed in the development by the Roman Catholic Church of the *Index Librorum Prohibitorum,* a list of proscribed books, the origins of which go back (in a primitive form) to the 5th century AD and which continued to have official sanction well into the 20th century. The most spectacular instance of the silencing of a thinker of note may well have been the restrictions placed upon Galileo in 1633.

The orthodoxy protected by an institution such as the *Index* probably had to be a system of thought in which much was made of certain books, particularly if other publications should seem to challenge in significant respects the teachings of the canonical texts. This must have appeared even more acute a problem when means became available, especially after the invention of printing, to produce and distribute books in large quantities.

The establishment of a fairly precise orthodoxy led to a perhaps unprecedented recourse to creeds. Thus, for example, the Nicene Creed was promulgated in AD 325. It was devised to fend off a heretical threat to Christian doctrine—and it led, partly because of a unilateral change in wording made by the Western church, to a schism that has continued since 1054 between Eastern Orthodoxy and Roman Catholicism.

Thus, it very much mattered which doctrines people were taught and what came to be believed—and this was largely determined, as it usually is, by the action of some authority, ecclesiastical or temporal. Similar developments can be seen in the Islāmic world down to this day, as well as among contemporary Marxists.

It is difficult to distinguish religious and nonreligious elements in some of the more celebrated controversies of the medieval Christian world, just as it is today among Islāmic peoples. The persecutions of witches—which ranged across much of Europe from the 14th to the 18th century and cost hundreds of thousands, if not millions, of lives—can be understood as due to various political, social, and psychic disturbances as well as to strictly religious differences.

The trials of Joan of Arc in France (1431) and of Thomas More in England (1535) are notorious illustrations of the difficulty in distinguishing religious from political differences. Indeed, it has been common, because of the experiences of the Middle Ages and of the Renaissance, to see the cause of political liberty as intimately related to the cause of religious liberty (and especially the liberty to do without religion).

Joan of Arc and Thomas More

The Enlightenment, beginning in the 17th century, attempted to purge Europe of the censorship that found political despotism allied with religious traditionalism. Alexis de Tocqueville was astonished to find in the United States, in the 1830s, that it was possible for ordinary men who stood for political freedom to be, and to remain, religiously devout. This was not the typical combination in the Europe of his day.

Even so, it should be recognized that the rigorous medieval theological–political regime against which moderns have rebelled did have at its core a principle that subjected the exercise of will (or sovereignty) to the test of wisdom. This principle, upon which the contemporary dedication to freedom of speech may ultimately depend, is reflected in Thomas Aquinas' insistence in *De veritate* (Q. 23), ". . . to say that justice depends simply upon the will is to say that the divine will does not proceed according to the order of wisdom, and that is blasphemous."

THE 17TH AND 18TH CENTURIES

The struggle against censorship in the Anglo-American world in the 17th and 18th centuries took two principal forms. There was the effort to keep government from reviewing, before publication, any manuscript, and there was the effort to keep government from penalizing, after publication, any text that expressed forbidden sentiments. (There were throughout the Western world developments with respect to these matters similar to those in Great Britain and the United States, but they usually occurred later.)

The effort to eliminate "previous restraints" (also known as prior restraints) in Great Britain and in America had its roots in English constitutional experience. Previous restraint (or licensing) came to be regarded as an inheritance of Roman Catholic practices. And so, when the Anglican successor to the Roman Catholic Church was disestablished by the Puritans, it was evidently something of a shock to John Milton to find Parliament reinstating licensing in 1643.

Milton's
"Areopagi-
ica"

Milton's "Areopagitica" (1644) has remained the classic statement of the arguments against censorship, particularly in the form of previous restraint. Milton conceded that criminal prosecutions might, perhaps even should, follow upon the publication of certain writings. He insisted, however, that such works must not be suppressed before publication.

Critical to Milton's position in support of freedom of the press is something that may not have been implicit in the traditional pre-Miltonian position against censorship—his confidence that truth, "in a free and open encounter," will be able to overcome error. Related to this opinion is the assurance that it is a positive good for mankind to be exposed to error; only in this way may virtue be tested, strengthened, and made adequate to the trials of earthly life. Milton cannot praise "a fugitive and cloister'd vertue." All this seems to rest upon a Christian view of the world: truth may indeed win out in its encounter with error, if the struggle continues long enough and if divine aid is thrown into the balance, as Milton seems to assume it will be; a man must not only act virtuously, but he must also personally choose to do so; he must be prepared to be exposed to alternatives, as inevitably he will be, and he must choose rightly if he is to merit and secure eternal salvation.

A reliance upon due process of law (which Milton in effect calls for) is the vital concession that the community can be led to make to reason: it provides a safeguard that must be so well established in times of calm and reflection that it is held to firmly, as a tenet of a common political faith, when the community is almost beside itself with passion. And, Milton might add if he were to use modern terminology, due process provides the ground rules for that free and open encounter in which truth may indeed prevail over error.

Black-
tone's
Commen-
aries

Thus, it is against the polemical background provided by Milton's "Areopagitica" that the abandonment of prepublication censorship in England in 1695 could be properly seen as a great victory for liberty of the press in Anglo-American constitutional history. And so, in 1765–69, William Blackstone could say about the English common law with respect to liberty of the press in his *Commentaries on the Laws of England:*

> The liberty of the press . . . consists in laying no *previous* restraints upon publications, and not in freedom from censure for criminal matter when published. Every freeman has an undoubted right to lay what sentiments he pleases before the public; to forbid this, is to destroy the freedom of the press: but if he publishes what is improper, mischievous, or illegal, he must take the consequence of his own temerity.

The next major step in the Anglo-American response to censorship problems may be seen in the First Amendment to the Constitution of the United States. That amendment, ratified in 1791, provides:

> Congress shall make no law respecting an establishment of religion, or prohibiting the free exercise thereof; or abridging the freedom of speech, or of the press; or the right of the people peaceably to assemble, and to petition the Government for a redress of grievances.

Similar provisions may be found in most of the state constitutions in the United States, although the connection between political and religious liberty is not always recognized to be as intimate as it is in the First Amendment.

Freedom
of speech

Such a guarantee of freedom of the press as is found in U.S. constitutional documents has long been understood to foreclose the possibility of previous restraints, thereby confirming the definition of "liberty of the press" found in Blackstone's *Commentaries.* A few scholars and jurists have gone so far as to suggest that the First Amendment and the state constitutional provisions do no more than limit restraints prior to publication, but it is difficult to bring the "freedom of speech" language (often found in the same guarantees) within this suggestion, since there never had been (and, in the nature of things, could never easily be) previous restraints upon what might be spoken (as distinguished from what might be printed). Rather, as indicated above, "freedom of speech" is modelled upon the British parliamentary privilege, a privilege that should be generally available when a people becomes sovereign in the fashion of the American people. (In addition, the celebrated case of John Peter Zenger [1735] had already established for Americans the principle that truth was a defense in seditious libel prosecutions, thus going beyond Blackstone's position in still another respect.)

The traditional parliamentary privilege—which is still guaranteed in the United States to members of Congress and to state legislators—can be considered virtually absolute in the protection it provides legislators against being held accountable "in any other place" for what they utter in a legislative body. The question remains, of course, as to precisely what kinds of matters may be discussed freely, and without fear of sanction, by citizens entitled to such protection as is provided by the First Amendment.

The old-fashioned answer was that the kind of discussion primarily protected by the First Amendment is that of citizens engaged in investigating and assessing the public business. Such protected discussion may be found in art, in moral and scientific inquiry, and in advertising, as well as in obvious political discourse. Thus, whatever is suppressed *because* of political differences is likely, in the circumstances, to be "political." Another way of putting this is to say that the crime of seditious libel is not consistent with the First Amendment. Particularly influential spokesmen for this position in the 20th century have been Alexander Meiklejohn (1872–1964), Hugo L. Black (1886–1971), Harry Kalven, Jr. (1914–74), and Malcolm P. Sharp (1897–1980). It is a position epitomized by its questioning of the constitutionality of the Sedition Act enacted by Congress in 1798.

The Sedition Act made criminal the publication of "any false, scandalous and malicious writing . . . against the government of the United States, or either House of Congress . . . or the President . . . with intent to defame [them] or to bring them . . . into contempt or disrepute." This act, which was allowed to lapse after two years, has been generally repudiated by American jurists and scholars. The U.S. credo in these matters may well be found in Thomas Jefferson's First Inaugural Address (1801), in which he said, "If there be any among us who would wish to dissolve this Union or to change its republican form, let them stand undisturbed as monuments of the safety with which error of opinion may be tolerated where reason is left free to combat it."

The First Amendment guarantee of freedom of speech and of the press was anticipated, in effect, by the provision of the Constitution (drafted in 1787) that "Treason against the United States shall consist only in levying War against them, or in adhering to their Enemies, giving them Aid and Comfort." This, too, has made it difficult to prosecute citizens for their criticism of government, something that had been much easier to do under an expansive definition of treason. The First Amendment guarantee was anticipated as well by the assumption evident in the Declaration of Independence that a people is always entitled to examine and to assess the doings of its governments, all with a view to being able to replace any government deemed upon due consideration to be unsatisfactory.

There can be no doubt that the extensive freedom of Americans to discuss political matters can lead to serious abuses. But it is generally recognized that the abuses resulting from censorship of such discussion—whether in the form of previous restraints or in the form of postpublication sanctions—are apt to be even more serious for a community.

Perhaps not as generally recognized is that considerable self-restraint is required, especially on the part of intellectuals, if the best possible use is to be made of free discussion in the circumstances of a people. A call for such restraint (or self-censorship), as well as for both public enlightenment and respect for organized religion, may be seen in George Washington's Farewell Address (1796). Similarly, Lord Macaulay could say of the 19th-century British press, "Foreigners who dare not print a word reflecting on the government under which they live, are at a loss to understand how it happens that the freest press in Europe is the most prudish."

MODERN PRACTICES

The Soviet system. The *Index,* which was abolished by the Roman Catholic Church in 1966, may be seen in another form in the Soviet Union, where there is a comprehensive system of supervision of manuscripts before publication. (Similar control, in varying degrees, is exercised in other countries with Marxist governments.) Such supervision, in the light of official Communist Party doctrines, is not limited to political discussions or to books and newspapers but seems to cover all kinds of subjects and all forms of publication, including broadcasts. This leads, in effect, to considerable self-censorship by authors seeking to be published in some form. Of course, the more "unreliable" authors are simply refused publication in the conventional places.

The comprehensive Soviet system has led to the development of a sophisticated mode of guarded expression in print, which is designed to conform with official proprieties even as signals are given to the more perceptive about impending political shifts. It has, as well, led to (or perhaps permitted the perpetuation of) such expedients as the transformation of the circus clown into a tolerated means for exhibiting and venting public exasperation with the regime. All in all, in such circumstances, the bearing of persecution on the art of writing can be seen.

There have been, in recent decades, periodic relaxations of control in the Soviet Union, but a pervasive and shameless control by the Communist Party oligarchy remains. A limited uncensored circulation is effected by privately copied manuscripts; a few authors have had their manuscripts published abroad. And then there are materials that are smuggled into the Soviet Union, as well as foreign radio broadcasts, which are "jammed" from time to time. But these exceptional modes provide uncensored information and discussion only to a very small number of Soviet citizens. The large body of citizens knows, by and large, only what the government has chosen to reveal, often being ignorant of critical conditions and developments in its own country (such as the serious illness of a Soviet leader) that are well known abroad.

This sort of control is justified as necessary for the protection of the state and the welfare of its citizenry. Some of the restrictions are designed to permit retention of information that is considered vital to national security; others are designed to keep citizens from being "misled," especially since a proper understanding of "dialectical materialism" is said to be necessary to determine what is relevant and what contributes to the health of the community and the well-being and moral soundness of citizens. Variations of the arguments used in the Soviet Union today may be found in medieval apologetics, in Confucian doctrines, in Plato's *Republic,* and in UNESCO proposals for "a new world information and communication order."

Orwell's *Nineteen Eighty-four*

How seriously such arguments are to be taken depends, in part, on whose interest the rulers exercising this control truly serve. George Orwell, in his novel *Nineteen Eighty-four* (1949), displayed a ruling class that was evidently drawn in large part from his study of the practices of the ruling party in the Soviet Union. The rulers in *Nineteen Eighty-four,* when they speak most frankly, disavow serving any interest but their own, whatever they may say publicly about national security and social progress. Such ruthless self-centredness on the part of rulers fits the traditional definition of tyranny. (It is not necessary to be concerned here with whether or not such a completely self-centred tyranny, which must be rare, would be likely to withstand determined opposition by men and women willing to sacrifice themselves.)

The portrayal in *Nineteen Eighty-four* is particularly gripping because it suggests how extensive the modern control of ideas can be. (Developments with respect to computers and information banks subsequent to the publication of Orwell's novel make the prospects of comprehensive control seem even more ominous.) The ultimate in censorship may be seen, in the novel, in a technology and an ideology that permit government to edit not only what is being said today and tomorrow but also what is recorded or remembered to have been said yesterday. The Western reader of *Nineteen Eighty-four* is likely to be offended not only by its government's efforts to control political discussion but also by the official assumption that the individual has no standing worthy of serious consideration.

It is unlikely that any system of censorship in the world today, with the possible exception of that in the People's Republic of China, is as effective as that in the Soviet Union. But official secrecy, as well as tyranny, is something to which Russians have long been accustomed. (In the late 19th century, the Russians had perhaps the only extensive censorship system in the world.) It is instructive, therefore, to consider how censorship works today in countries with somewhat more experience in self-government.

Censorship in Greece, 1967–74

Censorship under a military government. Particularly revealing in this respect was what happened in Greece between 1967 and 1974, when a conspiracy of junior army officers seized control of the government. The dependence of Greece upon foreign trade and tourism made it difficult to keep out the foreign press and foreign broadcasts. This meant, among other things, that the more educated citizens in the country were always fairly well informed about what was going on in the world at large. But information about domestic affairs (especially economic data) was hard to come by, since much of that kind of information depends in modern times (as in ancient China) upon official sources. (Thus, there is the better known example of the chronic complaint about the unreliability of official Soviet statistics. Thus, also, the strict censorship in Poland during the 1970s evidently kept the Communist government there from becoming aware of how serious the country's economic problems were, leading to considerable domestic turmoil. Such regimes depend, in effect, upon free peoples to do their thinking for them about the most serious matters.)

The limits of government censorship in a country such as Greece, where the press (unlike the broadcast media) is not owned by the government, are in part determined by the fact that much of the business of daily life depends on fairly reliable newspaper operations. All kinds of information—about goods for sale, about schedules and timetables, about innumerable activities upon which an efficient daily life depends—must be published regularly and reliably in the press, whoever may be in power. This means that newspapers must not be unduly delayed in their appearance; it also means that if they are to continue to appear they must be profitable.

The censor who is too slow (that is, careful) in reviewing everything that is to appear in a forthcoming daily newspaper jams up the works. And if he is too restrictive in what he does permit, the paper is apt to become so dull that readers do not buy it. Either way, sales suffer and newspapers go out of business.

What happens in practice is that a rough accommodation develops between an editor and his censor. Each can make the duties of the other a constant aggravation. The accommodation worked out is rather like that which guards and inmates arrive at in their collaborative governance of a prison. One critical problem in maintaining indefinitely a system of censorship is, as Milton pointed out, that it is dull, unrewarding work for the typical censor—and so the quality of people drawn to it tends to deteriorate.

Of course, one way of avoiding much of the difficulty, expense, and inefficiency of a system of prepublication censorship is simply to allow each editor to publish as he chooses, subject to the risk of prosecution for whatever is published contrary to the standards laid down by the regime. But it is far from easy, even in a dictatorial regime, to prosecute effectively so long as some semblance of due process remains. It appears simpler for dictators to refuse to permit a particular newspaper report to be published than it is to explain in open court what was wrong with the report once published. Whether it is indeed simpler can be doubted, however, considering the mammoth effort required to supervise the many innocuous reports that make up the bulk of any newspaper.

Two meanings of "censorship"

It should be evident from these observations that "censorship" is used today in two senses. The more limited, perhaps more rigorous, sense refers to a system of prepublication control; the broader sense includes, in addition, sanctions visited upon a publisher after publication (whether or not the publication has previously been "approved"). Something analogous to prepublication censorship is often said, by contemporary psychologists, to operate in the human psyche to prevent the conscious awareness of any unacceptable desires harboured in the unconscious. Comparable suppression, as well as intimidation, may be seen in the political world, when prosecution and persecution for various kinds of associations and actions can render certain opinions virtually unthinkable.

Postpublication censorship does tend to be moderated to the extent that there is the rule of law in the community (including trials that are conducted more or less in public). The Greek military government of 1967–74 was repeatedly embarrassed by the trials it dared to conduct in public. The same could be said of the South African government since World War II, so long as an independent judiciary was trying sedition cases. (One result of this was that certain cases involving "national security" were removed, by act of the South African Parliament, from the ordinary jurisdiction of the courts. Or, to put this in terms familiar in Anglo-American law, nothing comparable to a habeas corpus hearing is permitted in South Africa today in certain categories of cases.) In the Soviet Union, on the other hand, the judicial proceeding in a political case seems, by and large, to be but another tool of government policy: in such circumstances, there may not be much to choose from between prior restraint and postpublication sanctions if an efficient allocation of resources is not a concern.

The *Pentagon Papers*

Censorship in the United States. *Freedom of the press.* One of the most dramatic attempts by the government of the United States to exercise prepublication restraint occurred in connection with the *Pentagon Papers* (1971), a "top secret" multivolume report on the Vietnam War that was surreptitiously supplied to various newspapers, which then began to publish it in installments. Each newspaper that managed to secure and thereupon to publish the report was enjoined in turn, at the request of the U.S. Department of Justice. The Supreme Court of the United States, after hearing arguments, lifted the injunctions and publication proceeded.

This case points up how difficult it is, in the United States today, to prevent publication, whatever recourse there may be to criminal sanctions or to damage suits after unauthorized or improper publication. Of course, it cannot be known whether the Supreme Court would have acted differently if the *Pentagon Papers* documents had been more current or if they had dealt with even more sensitive materials.

By the very nature of things, prepublication restraint is, in the United States, a rare occurrence. Thus, if each newspaper that began to publish the *Pentagon Papers* had published in one issue everything it had, that would have been the end of the previous-restraint case. And it should be obvious that that is the typical situation in the United States: the first the government usually knows about any publication is when the newspaper comes out—and by that time, of course, prepublication restraint is out of the question.

Thus, the U.S. government, in order to keep certain information out of the press, has to depend upon its ability to select those to whom sensitive information may be entrusted. Otherwise, the judgment of editors must be relied upon. There is nothing in the United States comparable to the Official Secrets Act and the "D" notice system in Great Britain, which, it seems, effectively restrain editors from publishing materials bearing certain restrictive designations. An attempt was made in 1983–84 by the U.S. government to require thousands of officials handling classified matter to pledge that they would submit any future writings for prepublication review by government censors. Opposition in Congress kept the new code from taking effect, except perhaps to a limited degree.

Private property and censorship

Contributing massively to the absence of censorship in any country is the existence there of considerable private property. Not only do personal resources provide a cushion against government unfriendliness, but they also provide independent access to the means of publication, if only in the form of a private printing with private circulation or of paid advertisements in the press. The attempts at censorship encountered in the United States today testify, in effect, to the importance of private property for freedom of the press. The instances of censorship that have been widely publicized generally have to do with public libraries, textbook selections, and government employment contracts. But in these instances, as with most of the repressive measures of the 1940s and 1950s, public funding, government authority, or a critical dependence upon public opinion—as in the case of the motion-picture and broadcast industries—is involved. Otherwise, there would be no effective way for either the government or public opinion to control what is published—certainly not when anyone with private means is determined to make his opinions known.

Parallel to the immunity provided by the institution of private property is that provided in the United States by academic freedom in colleges and universities. This freedom, which encourages scholars and teachers to traffic in unpopular truths, rests in part on the private property of tenured appointments. On the other hand, libel suits on behalf of another kind of private property—the interest everyone has in his reputation—are seen by some as a growing danger to freedom of the press. That is, concern has been expressed lest the protection provided by *New York Times Co.* v. *Sullivan* (1964) be eroded. In that case, the U.S. Supreme Court required that any public official who sues for damages because of an alleged falsehood prove that the falsehood had been issued with knowledge that it was false or in reckless disregard of whether it was false or not. The court was determined to protect the press from the prospects of large damage awards in libel cases that would intimidate it into drastic self-censorship. The court also saw itself as confirming the settled U.S. opinion condemning the Sedition Act of 1798.

A different kind of "protection" for the press, less welcome to journalists, was the decision by the U.S. government not to permit reporters to accompany the troops invading Grenada in 1983. Critical to the controversy on that occasion was a general concern that too much of the information necessary for adequate discussion of public affairs remains within the exclusive control of government. Thus, it is sometimes said, neither previous restraints nor postpublication sanctions need be resorted to by a government able to shape public opinion simply by regulating the flow of vital information as it pleases. This, too, can be considered a form of censorship, the more insidious in that it is obviously sensible in some cases to restrict public access to information for the sake of legitimate defense, diplomatic and administrative efficiency, or confidential professional relations.

Still another form of censorship may take the form of the preferences government bodies exhibit through the financial and other support they distribute to artistic, scientific, and educational applicants. And yet it is generally recognized that such distribution can be helpful, perhaps even necessary, and that it has to be done on the basis of standards that must rely on the good-faith judgment of public officials for their application. Here, as elsewhere, an informed and vigilant citizenry may be the best guarantor of both quality and fairness.

Freedom and truth. Postpublication sanctions were used in the courts, between 1948 and 1961, against leaders of the Communist Party in the United States. Even so, the indictments in those cases were put in terms of a conspiracy to overthrow the government. That is, despite the unpopularity of Communism in a time of considerable international tension, no U.S. government could rely merely on the fact that people found the defendants' opinions to be offensive. An effort had to be made to connect what the defendants were saying to what they (and others elsewhere) were likely to do.

Still, such prosecutions were confronted by the prohibition in the First Amendment that "Congress shall make no law . . . abridging the freedom of speech, or of the press." But the apparent absoluteness of that prohibition had long been subverted by the ill-conceived, yet all too influential, statement by Justice Oliver Wendell Holmes in *United States* v. *Schenck* (1919):

> . . . the character of every act depends upon the circumstances in which it is done. The most stringent protection of free speech would not protect a man in falsely shouting fire in a theatre and causing a panic. [The] question in every case is whether the words used are used in such circumstances and are of such a nature as to create a clear and present danger that they will bring about the substantive evils that Congress has a right to prevent.

Dennis v. *United States*

There does not seem to be much doubt that the man who causes a panic in a theatre should be dealt with firmly. But it is far from clear that this sensible conclusion has justified punishing men and women whose principal offense seems to have been that of raising fundamental (however ill-conceived) objections to the established political, economic, and social arrangements in the United States. Justice Holmes's constitutional flexibility in the Schenck case can be considered to have culminated in the later assurance by Chief Justice Fred M. Vinson in *Dennis* v. *United States* (1951), in which the convictions of a dozen Communist Party leaders were upheld:

> Nothing is more certain in modern society than the principle that there are no absolutes, that a name, a phrase, a standard has meaning only when associated with the considerations which gave birth to the nomenclature. . . . To those who would paralyze our government in the face of impending threats by encasing it in a semantic straitjacket we must reply that all concepts are relative.

This is hardly in the spirit of Milton's high confidence in the power of an enduring truth to prevail. Nor is it in the spirit of the Declaration of Independence, with its informed reliance upon natural rights, upon self-evident truths, and hence upon the right of revolution.

Be that as it may, it is unlikely that any of the prosecutions from the Schenck to the Dennis case and beyond for what was, in effect, sedition would succeed under present conditions. The sorts of things those defendants said are no longer considered dangerous by the community at large. Rather, the much more vexing question is whether any kind of speech is not entitled to First Amendment protection. That protection is now said to extend far beyond political discussion. Thus, advertising (or commercial speech) is said to be protected as is much (if not virtually all) obscenity, although reservations are heard about child pornography and about inducements to violence and the worst depravity. (Whether any particular utterance or action should be regulated has itself always been a political question open to free discussion.)

Much is made today of an asserted right of self-expression and of the related right to privacy. The arguments drawn upon in their support seem to be variations of those developed in John Stuart Mill's *On Liberty* (1859). Mill's arguments are invoked today not only in opposition to government censorship but in opposition as well to those suppressive efforts by private organizations that are sometimes more effective than government can be in a liberal democracy. Particularly susceptible to the influence of private censors are the broadcasting media, especially since they are still subject in the United States to some government regulation. A different kind of private suppression has been usefully described in this fashion by Jamie Kalven: "Being badly edited is as close as most American writers ever come to being censored. It thus offers a vehicle for imagining the *experience* of censorship, for getting at what it *feels* like. My strongest impression is that the abuse of one's prose feels like an assault on one's mind." Similarly, Lord Radcliffe (in his *Censors*) could speak of "the real licensors of thought today, the editors, the publishers, the producers, the controllers of radio and television."

Character and freedom

Lord Radcliffe could speak as well about "the apparent indifference of censors of all kinds to the depiction or portrayal of mindless violence and brutality, that witless rejection of civility that threatens to be the Black Death of the twentieth century." Thus, it is not usually noticed today that Mill recognized that a people has to be trained properly to make use of the considerable liberty he advocates. If, for example, a community should recognize that television is corrupting the young, distorting the political process, and generally playing havoc with education and the public character, is it really helpless to do anything about it? Would it be censorship to abolish altogether such a baleful influence? And if abolition of television should be considered censorship, may not that suggest that censorship is not altogether bad? What, in short, is the popular character presupposed for effective self-government, and how is that character properly to be developed and maintained?

Such questions reflect the fact that censorship and freedom of the press problems depend for their sensible resolution upon more general considerations of liberty, of the common good, and of the rights, virtues, and duties of citizens entrusted with self-government. Thus, Tocqueville could observe in *Democracy in America* (1835–40):

> It cannot be repeated too often: nothing is more fertile in marvels than the art of being free, but nothing is harder than freedom's apprenticeship. The same is not true of despotism. Despotism often presents itself as the repairer of all the ills suffered, the support of just rights, defender of the oppressed, and founder of order. People are lulled to sleep by the temporary prosperity it engenders, and when they do wake up, they are wretched. But liberty is generally born in stormy weather, growing with difficulty amid civil discords, and only when it is already old does one see the blessings it has brought.

Among the blessings of liberty may be found the philosophical pursuits that have sometimes appeared so threatening to public order. Laurence Berns has reformulated the ancient dilemma posed by the trial of Socrates, "the greatest hero of freedom of thought"—a dilemma that exposes one of the roots of the perennial censorship controversy:

> Is philosophy, the intransigent quest for truth (including the truth about politics and religion), inherently subversive? Does it necessarily undermine political society and conventional morality, or, on the contrary, is a good society impossible without freedom to philosophize?

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(G.An.)

Central Africa

Although merely a conventional designation, the term central Africa corresponds with an undeniable physiographic whole, consisting of the interior bowl of the Congo River basin and the intervening spaces that lie between it and the Atlantic Ocean. Stretching across the African continent from latitude 11° N to 13° S and from longitude 8° 43′ E (at Cape Lopez) to 31° 30′ E (at Lake Albert), central Africa comprises Burundi, the Central African Republic, Congo, Gabon, Rwanda, and Zaire; it also includes the islands of São Tomé and Príncipe off the coast of Gabon. This is the least populated part of tropical Africa, with the exception of Rwanda and Burundi, which are the most densely populated countries on the continent. Rwanda and Burundi's human geography resembles that of eastern Africa, but they are here linked with central Africa because, before independence, they were administered by Belgium, as was neighbouring Zaire.

This article is divided into the following sections:

THE REGION

Physical and human geography

THE LAND

Geology. Central Africa is underlain by a basement complex of Precambrian rock, which is partially covered by a thin sedimentary layer. This complex consists of a varied lithology, including migmatites and other metamorphic rocks, granito-gneissic complexes, and partially metamorphosed sedimentary rocks (among which are limestones, sandstones, and conglomerates of glacial origin).

Orogeny

Several orogenic phases occurred more than 600 million years ago. Though long subjected to weathering and erosion, the folds produced by these orogenies are still obvious in some reliefs: ridges lie parallel to the coast in western Congo and coastal Zaire; linear and curvilinear ridges of quartzitic sandstones are found in Rwanda and Burundi; and an arc-shaped range extends through the copper belt of southeastern Zaire. This Precambrian complex contains a wealth of mineral resources: gold and manganese are found in the very ancient parts (northeastern and southern Zaire, southeastern Gabon) and tin in Middle Precambrian strata (southeastern Zaire, Rwanda, Burundi), and in Late Precambrian strata (western Congo, southeastern Zaire) there is copper, often linked with lead and zinc or, as in southern Zaire, with cobalt and uranium. Iron ore abounds in central Africa's crust, but it is still relatively unexploited. During the Cretaceous Period (144 to 66.4 million years ago), kimberlitic pipes, a major source of diamonds, intruded through the basement rocks to the surface in central Zaire.

In the west the basement sinks under a sedimentary coastal basin formed during the Cretaceous Period—*i.e.,* after Africa and South America were separated. Narrow in Zaire, the basin widens in Gabon. Petroleum (most often tapped offshore), some phosphate, and potash are found there.

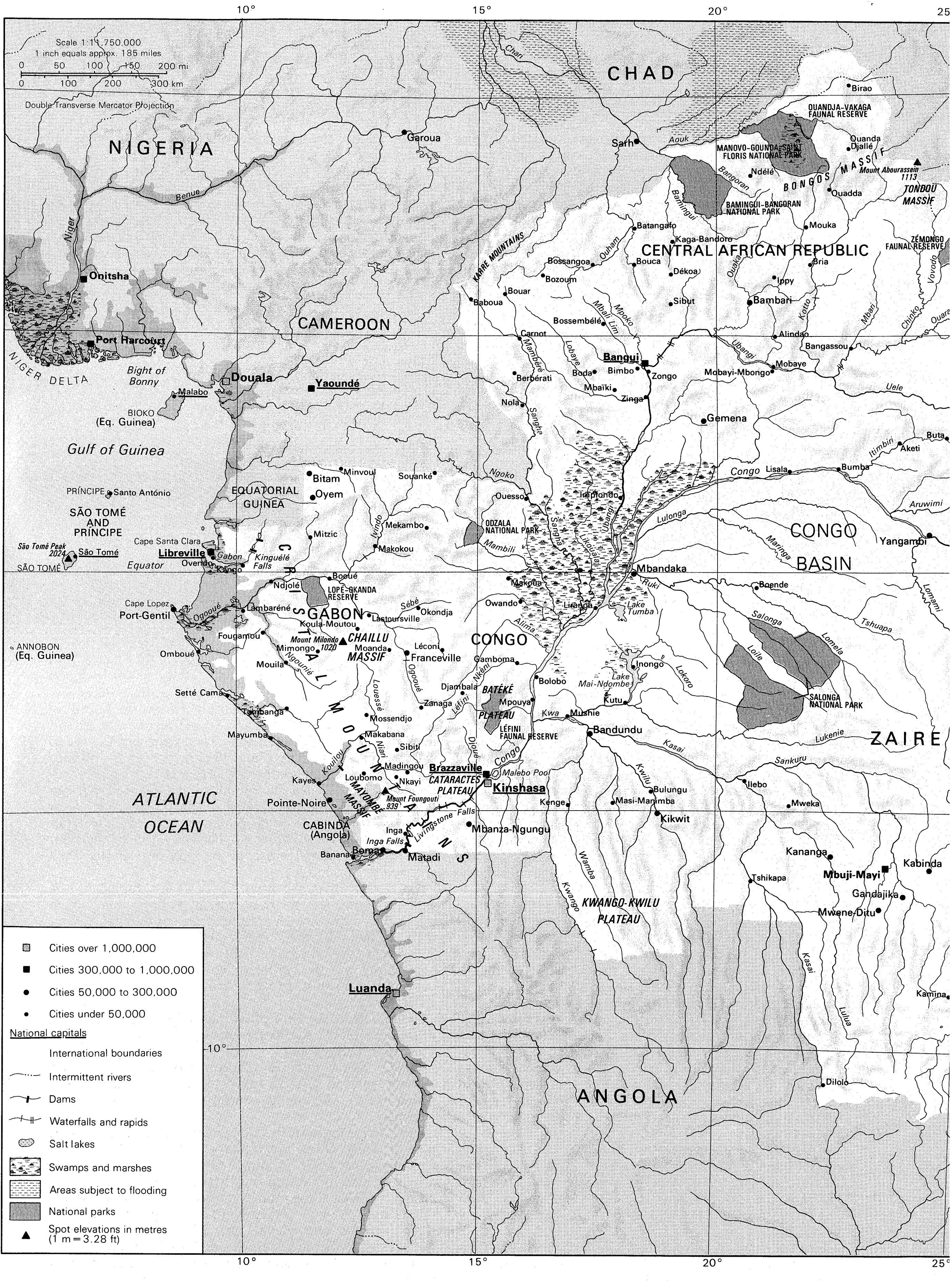
Since the end of the Precambrian, the basement complex has undergone deformations of major amplitude, in addition to local upliftings and fractures. Raised and emerging at the periphery, the basement sags in the centre into a shallow basin (maximum depth about 5,000 to 6,500 feet [1,500 to 2,000 metres]), buried beneath sediments—often of continental origin—deposited from the Late Precambrian to the Quaternary. The sediments also extend toward the south, between the coastal rim and the Kasai region in western Zaire. Sand deposits cover the surface of some plateaus in southern Zaire, and other small sedimentary basins may be found in the Central African Republic (*e.g.,* sandstone plateaus near Carnot in the west and between Ouadda and Mouka in the east, the latter containing alluvial diamonds). The far north of the Central African Republic is part of the interior-draining Chad sedimentary basin.

The eastern part of central Africa is interrupted by the western branch of the East African Rift System, which marks the progressive separation of the continental African plate in the west from the Somalian plate in the east. Transform faults disconnect the valley axis. Two of these dislocations are the centre of volcanic activity (old volcanism in the south, present Virunga volcanism north of Lake Kivu).

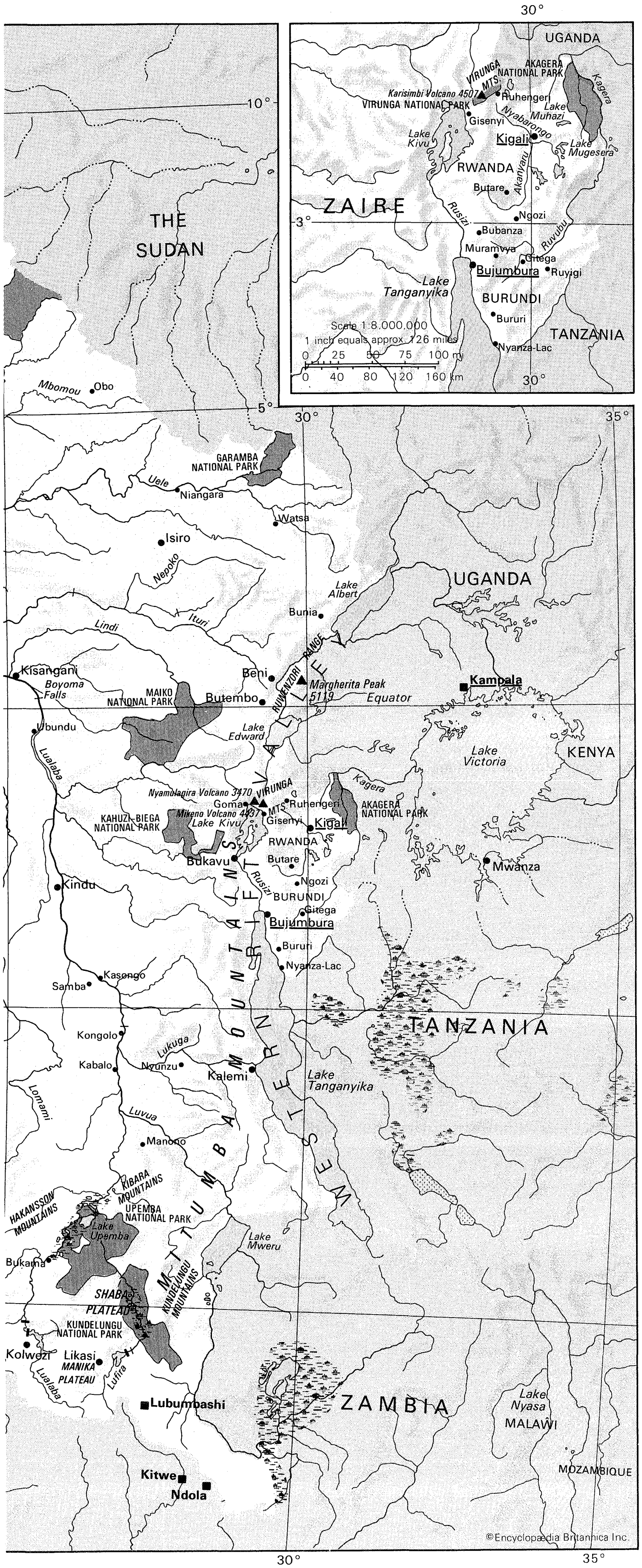
The islands in the Gulf of Guinea are of volcanic origin, jutting out of the sea along an extension of the western Cameroon fault axis.

Relief. The landscapes of central Africa are most often wide plateau surfaces (often terraced planed surfaces), which are smooth in the central part and etched at the periphery.

The interior basin of the Congo (Zaire) River, the lowest part of which (Malebo [Stanley] Pool) lies about 1,000 feet above sea level, is joined to the Atlantic Ocean by a narrow neck traversing ridges parallel to the coast. The basin comprises some marshy surfaces in the region where the Congo, Ubangi (Oubangui), Likouala, and Sangha rivers converge and where Lakes Mai-Ndombe (formerly Lake Léopold II) and Tumba are found. Its major part, however, consists of drier surfaces (low plateaus or alluvial terraces).



▨	Cities over 1,000,000
■	Cities 300,000 to 1,000,000
●	Cities 50,000 to 300,000
•	Cities under 50,000
National capitals	
	International boundaries
	Intermittent rivers
	Dams
	Waterfalls and rapids
	Salt lakes
▨	Swamps and marshes
▨	Areas subject to flooding
▨	National parks
▲	Spot elevations in metres (1 m = 3.28 ft)



MAP INDEX

Cities and towns

Aketi	2 44 N	23 46 E
Albertville, see Kalemi		
Alindao	5 02 N	21 13 E
Baboua	5 48 N	14 49 E
Bambari	5 45 N	20 40 E
Banana	6 01 S	12 24 E
Bandundu	3 19 S	17 22 E
Bangassou	4 44 N	22 49 E
Bangui	4 22 N	18 35 E
Batangafo	7 18 N	18 18 E
Beni	0 30 N	29 28 E
Berbérati	4 16 N	15 47 E
Bimbo	4 18 N	18 33 E
Birao	10 17 N	22 47 E
Bitam	2 05 N	11 29 E
Boda	4 19 N	17 28 E
Boende	0 13 S	20 52 E
Bolobo	2 10 S	16 14 E
Boma	5 51 S	13 03 E
Booué	0 06 S	11 56 E
Bossangoa	6 29 N	17 27 E
Bossembélé	5 16 N	17 39 E
Bouar	5 57 N	15 36 E
Bouca	6 30 N	18 17 E
Bozoum	6 19 N	16 23 E
Brazzaville	4 16 S	15 17 E
Bria	6 32 N	21 59 E
Bubanza	3 06 S	29 23 E
Bujumbura	3 23 S	29 22 E
Bukama	9 12 S	25 51 E
Bukavu	2 30 S	28 52 E
Bulungu	4 33 S	18 36 E
Bumba	2 11 N	22 28 E
Bunia	1 34 N	30 15 E
Bururi	3 57 S	29 37 E
Buta	2 48 N	24 44 E
Butare	2 36 S	29 44 E
Butembo	0 09 N	29 17 E
Carnot	4 56 N	15 52 E
Dékoa	6 19 N	19 04 E
Dilolo	10 42 S	22 20 E
Djambala	2 33 S	14 45 E
Elisabethville, see Lubumbashi		
Fougamou	1 13 S	10 36 E
Franceville	1 38 S	13 35 E
Gamboma	1 53 S	15 51 E
Gandajika	6 45 S	23 57 E
Gemena	3 15 N	19 46 E
Gisenyi	1 42 S	29 15 E
Gitega	3 26 S	29 56 E
Goma	1 41 S	29 14 E
Ilebo	4 19 S	20 35 E
Impfondo	1 37 N	18 04 E
Inga	5 39 S	13 39 E
Inongo	1 57 S	18 16 E
Ippy	6 15 N	21 12 E
Isiro	2 46 N	27 37 E
Kabalo	6 03 S	26 55 E
Kabinda	6 08 S	24 29 E
Kaga Bandoro	6 59 N	19 11 E
Kalemi (Albertville)	5 56 S	29 12 E
Kamina	8 44 S	25 00 E
Kananga (Luluabourg)	5 54 S	22 25 E
Kango	0 09 N	10 08 E
Kasongo	4 27 S	26 40 E
Kayes	4 25 S	11 41 E
Kenge	4 52 S	16 59 E
Kigali	1 57 S	30 04 E
Kikwit	5 02 S	18 49 E
Kindu	2 57 S	25 56 E
Kinshasa (Léopoldville)	4 18 S	15 18 E
Kisangani (Stanleyville)	0 30 N	25 12 E
Kolwezi	10 43 S	25 28 E
Kongolo	5 23 S	27 00 E
Koula-Moutou	1 08 S	12 29 E
Kutu	2 44 S	18 09 E
Lambaréné	0 42 S	10 13 E
Lastoursville	0 49 S	12 42 E
Léconi	1 35 S	14 14 E
Léopoldville, see Kinshasa		
Libreville	0 23 S	9 27 E
Likasi	10 59 S	26 44 E
Liranga	0 40 S	17 36 E
Lisala	2 09 N	21 31 E
Loubomo	4 12 S	12 41 E
Lubumbashi (Elisabethville)	11 40 S	27 28 E
Luluabourg, see Kananga		
Madingou	4 09 S	13 34 E
Makabana	2 48 S	12 29 E
Makokou	0 34 N	12 52 E
Makoua	0 01 N	15 39 E
Manono	7 18 S	27 25 E
Masi-Manimba	4 46 S	17 55 E
Matadi	5 49 S	13 27 E
Mayumba	3 25 S	10 39 E
Mbaïki	3 53 N	18 00 E
Mbandaka	0 04 N	18 16 E
Mbanza-Ngungu	5 15 S	14 52 E
Mbuji-Mayi	6 09 S	23 36 E
Mekambo	1 01 N	13 56 E
Mimongo	1 38 S	11 39 E
Minvoul	2 09 S	12 08 E
Mitzic	0 47 N	11 34 E
Moanda	1 34 S	13 12 E
Mobaye	4 19 N	21 11 E
Mobayi-Mbongo	4 18 N	21 11 E
Mossendjo	2 57 S	12 44 E
Mouila	1 52 S	11 01 E
Mouka	7 16 N	21 52 E
Mpouya	2 37 S	16 13 E
Muramvya	3 16 S	29 37 E
Mushie	3 01 S	16 54 E
Mweka	4 51 S	21 34 E
Mwene-Ditu	7 03 S	23 27 E
Ndélé	8 24 N	20 39 E
Ndjolé	0 11 S	10 45 E
Ngozi	2 54 S	29 50 E
Niangara	3 42 N	27 52 E
Nkayi	4 11 S	13 18 E
Nola	3 32 N	16 04 E
Nyanza-Lac	4 21 S	29 36 E
Nyunzu	5 57 S	28 01 E
Obo	5 24 N	26 30 E
Okondja	0 41 S	13 47 E
Omboué	1 34 S	9 15 E
Ouadda	8 04 N	22 24 E
Ouanda Djallé	8 54 N	22 48 E
Ouesso	1 37 N	16 04 E
Ovendo	0 17 N	9 30 E
Owando	0 29 S	15 55 E
Oyem	1 37 N	11 35 E
Pointe-Noire	4 48 S	11 51 E
Port-Gentil	0 43 S	8 47 E
Ruhengeri	1 30 S	29 38 E
Ruyigi	3 29 S	30 15 E
Samba	4 38 S	26 22 E
Santo António (São Antõnio)	1 39 N	7 26 E
São Tomé	0 20 N	6 44 E
Setté Cama	2 32 S	9 45 E
Sibiti	3 41 S	13 21 E
Sibut	5 44 N	19 05 E
Souanké	2 05 N	14 03 E
Stanleyville, see Kisangani		
Tchibanga	2 51 S	11 02 E
Tshikapa	6 25 S	20 48 E
Ubundu	0 21 S	25 29 E
Watsa	3 03 N	29 32 E
Yangambi	0 47 N	24 28 E
Zanaga	2 15 S	13 50 E
Zinga	3 43 N	18 35 E
Zongo	4 21 N	18 36 E

Physical features and points of interest

Abourassein, Mount	8 39 N	24 14 E
Akagera, see Kagera		
Akagera National Park	1 30 S	30 35 E
Akanyaru, *river*	2 05 S	30 01 E
Albert (Mobutu Sese Seko), Lake	1 40 N	31 00 E
Alima, *river*	1 36 S	16 36 E
Aouk, *river*	8 51 N	18 53 E
Aruwimi, *river*	1 13 N	23 36 E
Atlantic Ocean	5 00 S	10 00 E
Bamingui, *river*	8 33 N	19 05 E
Bamingui-Bangoran National Park	8 00 N	20 00 E
Bangoran, *river*	8 42 N	19 06 E
Batéké Plateau	3 30 S	15 45 E

Bomu, see Mbomou		
Bongos (Mongos) Massif	8 40 N	22 25 E
Boyoma Falls	0 30 N	25 12 E
Cataractes Plateau	4 27 S	13 52 E
Chaillu Massif	2 32 S	11 10 E
Chinko, *river*	4 50 N	25 53 E
Congo (Zaire), *river*	6 04 S	12 24 E
Congo Basin	0 00 N	20 00 E
Cristal Mountains	5 30 S	13 30 E
Djoué, *river*	4 19 S	15 14 E
Edward (Idi Amin), Lake	0 21 S	29 35 E
Foungouti, Mount	4 23 S	12 43 E
Gabon, *estuary*	0 25 N	9 20 E
Garamba National Park	4 10 N	29 30 E
Guinea, Gulf of	2 00 N	5 00 E
Hakansson Mountains	8 50 S	25 40 E
Idi Amin, see Edward, Lake		
Inga Falls	5 45 S	13 30 E
Itimbiri, *river*	2 02 N	22 44 E
Ituri, *river*	1 40 N	27 01 E
Ivindo, *river*	1 44 S	11 37 E
Kagera (Akagera), *river*	0 57 S	31 47 E
Kahuzi-Biega National Park	1 50 N	28 40 E
Karisimbi Volcano	1 30 S	29 27 E
Karre Mountains	6 33 N	15 40 E
Kasai, *river*	3 02 S	16 57 E
Kibara Mountains	8 35 S	27 25 E
Kinguélé Falls	0 29 N	10 20 E
Kivu, Lake	2 00 S	29 10 E
Kotto, *river*	4 14 N	22 02 E
Kouilou, *river*	4 28 S	11 41 E
Kundelungu Mountains	10 00 S	27 55 E
Kundelungu National Park	10 30 S	27 45 E
Kwa, *river*	3 10 S	16 11 E
Kwango, *river*	3 14 S	17 22 E
Kwango-Kwilu Plateau	6 00 S	18 00 E
Kwilu, *river*	3 22 S	17 22 E
Léfini, *river*	2 57 S	16 10 E
Léfini Faunal Reserve	2 58 S	15 25 E
Léopold II, see Mai-Ndombe, Lake		
Likouala, *river*	0 50 S	17 11 E
Lindi, *river*	0 33 N	25 05 E
Livingstone Falls	4 50 S	14 30 E
Lobaye, *river*	3 41 N	18 35 E
Loile, *river*	1 21 S	22 13 E
Lokoro, *river*	1 43 S	18 23 E
Lomami, *river*	0 46 N	24 16 E
Lomela, *river*	0 14 S	20 42 E
Lopé-Okanda Reserve	0 30 S	11 30 E
Lopez, Cape	0 37 S	8 43 E
Louessé, *river*	3 25 S	12 38 E
Lualaba, *river*	3 01 S	26 56 E
Lufira, *river*	8 16 S	26 27 E
Lukenie, *river*	2 44 S	18 09 E
Lukuga, *river*	5 40 S	26 55 E
Lulonga, *river*	0 43 N	18 23 E
Lulua, *river*	5 02 S	21 07 E
Luvua, *river*	6 46 S	26 58 E
Mai-Ndombe (Léopold II), Lake	2 00 S	18 20 E
Maiko National Park	3 30 N	27 45 E
Malebo (Stanley) Pool	4 15 S	15 25 E
Mambéré, *river*	3 31 N	16 03 E
Mambili, *river*	0 07 N	16 08 E
Manika Plateau	10 00 S	26 00 E
Manovo-Gounda-Saint Floris National Park	9 00 N	22 00 E
Margherita (Marguerite) Peak	0 22 N	29 51 E
Maringa, *river*	1 14 N	19 48 E
Mayombé Massif	4 30 S	12 30 E
Mbali Lim (Mbali), *river*	4 27 N	18 20 E
Mbari, *river*	4 34 N	22 43 E
Mbomou (Bomu), *river*	4 08 N	22 27 E
Mikeno Volcano	1 25 S	29 25 E
Milondo, Mount	1 43 S	12 06 E
Mitumba Mountains	6 00 S	29 00 E
Mobutu Sese Seko, see Albert, Lake		
Moero, see Mweru, Lake		
Mongos, see Bongos Massif		
Mpoko, *river*	4 19 N	18 33 E
Mugesera, Lake	2 06 S	30 20 E
Muhazi (Mohasi), Lake	1 51 S	30 24 E
Mweru (Moero), Lake	9 00 S	28 45 E
Nepoko, *river*	1 40 N	27 01 E
Ngounié, *river*	0 37 S	10 18 E
Niari, *river*	3 56 S	12 12 E
Nkéni, *river*	2 00 S	16 19 E
Nyabarongo, *river*	2 21 S	30 22 E
Nyamulagira Volcano	1 25 S	29 12 E
Odzala National Park	1 00 S	15 00 E
Ogooué, *river*	0 49 S	9 00 E
Ouaka, *river*	4 59 N	19 56 E
Ouandja-Vakaga Faunal Reserve	9 15 N	22 10 E
Ouara, *river*	5 59 N	20 14 E
Oubangui, see Ubangi		
Ouham, *river*	9 18 N	18 14 E
Príncipe, *island*	1 37 N	7 25 E
Ruki, *river*	0 05 N	18 17 E
Rusizi (Ruzizi), *river*	2 23 S	29 13 E
Ruvubu, *river*	2 23 S	30 47 E
Ruwenzori Range	0 23 N	29 54 E
Salonga, *river*	0 10 S	19 50 E
Salonga National Park	1 45 S	21 20 E
Sangha, *river*	1 13 S	16 49 E
Sankuru, *river*	4 17 S	20 25 E
Santa Clara, Cape	0 30 N	9 19 E
São Tomé, *island*	0 12 N	6 39 E
São Tomé Peak	0 16 N	6 33 E
Sébé, *river*	1 02 S	13 06 E
Shaba Plateau	10 00 S	27 00 E
Stanley, see Malebo Pool		
Tanganyika, Lake	6 00 S	29 30 E
Tondou Massif	7 50 N	23 45 E
Tshuapa, *river*	0 14 S	20 42 E
Tumba, Lake	0 48 S	18 03 E
Ubangi (Oubangui), *river*	0 30 S	17 42 E
Uele, *river*	4 09 N	22 26 E
Upemba, Lake	8 36 S	26 26 E
Upemba National Park	9 00 S	26 00 E
Virunga Mountains	1 00 S	29 50 F
Virunga National Park	1 00 S	29 15 E
Vovodo, *river*	5 40 N	24 21 E
Wamba, *river*	3 56 S	17 12 E
Western Rift Valley	3 00 S	29 00 E
Zaire, see Congo		
Zémongo Faunal Reserve	6 45 N	25 15 E

Plateau formations

Higher plateaus extend through older sedimentary strata around the central part of the Congo basin. Those crossed by the Congo above Malebo Pool—the Batéké Plateaus—reach an elevation of 2,600 to 3,000 feet north of Brazzaville and stretch southeastward to the Kwango Plateaus, which exceed 3,000 feet near the Angolan border. Beyond this sedimentary aureole the plateaus frequently show outcroppings of Precambrian basement rock. In the north a low divide (2,000–2,300 feet) separates the Congo River and its tributaries from the Chad basin. The landscape beyond the divide descends by steps toward Lake Chad. In the west the rim of the Congo basin is well demarcated. The Chaillu Massif (3,000–3,300 feet), in Gabon and Congo, is a granitic elliptical buttress surrounded by a crown of ancient erosion surfaces. Southwest and south of the massif, from west to east, are ridges ranging from 2,600 to 3,000 feet, which are traversed through deep and narrow gorges by the Kouilou and Congo rivers; a shale-limestone depression; and a sandstone plateau. East of this complex the Congo River has eroded a broad basin, known as Malebo Pool, into the upper sedimentary strata before cutting rapids farther downstream. In the southeast, the plateaus often exceed 3,300 feet. Some parts, uplifted more than 5,000 feet (*e.g.,* Kibara, Kundelungu), surround depressions (*e.g.,* Lake Upemba, Lufira River plain).

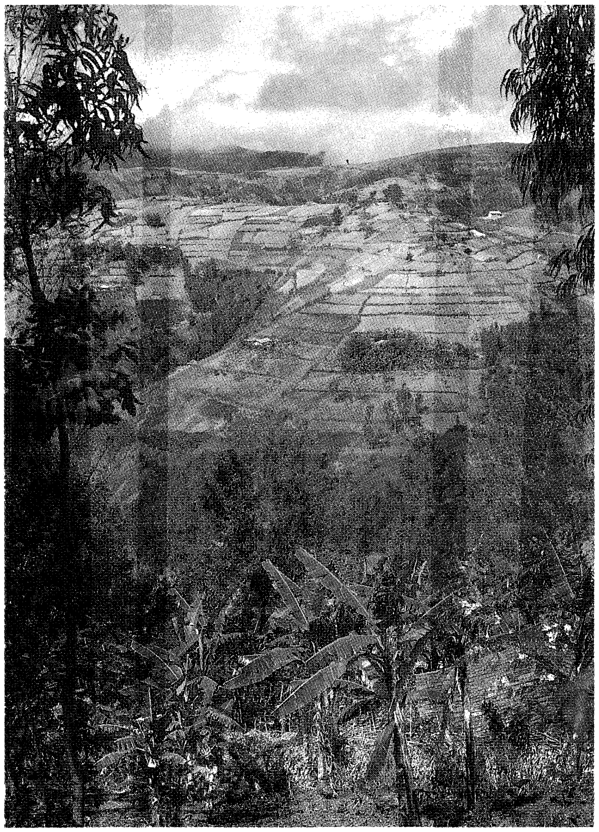
The most rugged terrain is seen on the eastern fringe of the Congo basin, where the Precambrian bedrock has been uplifted above the Western Rift Valley and is more than 8,000 feet high. A ravined slope abruptly drops to the floor of the rift valley. The level of Lake Albert (Mobutu) lies at 2,034 feet, that of Lake Edward at 2,995 feet, and that of Lake Tanganyika at 2,543 feet. In its southern part, the bottom of Lake Tanganyika is 2,165 feet below sea level, making the lake more than 4,700 feet deep. In Rwanda and Burundi, which lie east of the Western Rift Valley, the eastern edge of the valley mirrors that of the west: a vigorous slope on the fault side and an etched surface with a gentler slope dipping toward the east. North of Lake Kivu and of Rwanda, the Virunga volcanoes form an east–west-trending range whose main peaks exceed 9,800 feet (14,557 feet [4,437 metres] for Mikeno, 14,787 feet [4,507 metres] for Karisimbi). The highest point in central Africa, Margherita Peak (16,795 feet [5,119 metres]), is located on an uplifted portion of the Precambrian basement on the eastern fringe of the Rift Valley. Its summit bears residual features of glaciation.

The western sedimentary basin (a maritime plain with low hills) has a coastline broken up by estuaries (that of the Congo, downstream from Boma, Zaire, extends into a deep submarine canyon), lagoons, and rias. In Congo and southern Gabon offshore sandbars, stretched toward the north by the coastal current, alternate with small sea cliffs or rocky capes.

Drainage. Central Africa encompasses the Congo River basin. The immense size of the basin (1,334,700 square miles [3,457,000 square kilometres]) and its extension into regions with heavy rainfall make this river second only to the Amazon for rate of flow, which averages 1,421,400 cubic feet per second (40,250 cubic metres per second) at Kinshasa.

The Congo's flow fluctuations during the year, important in absolute value, are moderate in relative value: the low water average is 1,094,610 cubic feet per second, and the high water average is 1,977,360 cubic feet per second. Because the Congo straddles the equator, with its contrasting north-south precipitation regimes, the low waters of affluents originating in one hemisphere are matched by simultaneous high waters coming from affluents in the opposite hemisphere; a near constant flow is fed by the affluents of the central part. As the basin extends mostly south rather than north of the equator, the lowest waters occur in July and August—the Southern Hemisphere dry season—the highest waters in December. The tributaries in the peripheral regions have more varied regimes. As is true of the Amazon, the rivers rising in the forest are black, laden with humic acids (the name of the Mai-Ndombe, for example, means "black water"), whereas the rivers rising in the basin periphery are of brown or reddish brown waters.

In the central part of the basin the fan of quiet rivers constitutes one of the most attractive networks of navigable waters in the world, but this network is cut off from



Small farms line the slopes in the highlands of Burundi, one of the most densely populated regions in central Africa.

Dr. Nigel Smith/The Hutchison Library

the Atlantic by a succession of rapids between Kinshasa and Matadi, Zaire. Downstream from Matadi the river becomes navigable again before entering its estuary. All the rivers in the region flow down through rapids or waterfalls from the peripheral plateaus to the central basin. The rivers of the western region are only navigable for a few miles from their estuaries, though the Ogooué, in Gabon, which lies in a wider sedimentary coastal basin, is navigable for more than 100 miles.

Navigable rivers

Some parts of central Africa lie within the Chad basin (northern Central African Republic) or the Nile basin (eastern Rwanda and northeastern Burundi, the Rift Valley north of the Virunga Mountains). Lava fields from the Virungas have blocked the outlet of an ancient hydrographic network previously oriented to the north. This created Lake Kivu; its waters discharge to the south, through the Ruzizi (Rusizi) River gorge into Lake Tanganyika, one outlet of which links the lake with the Lualaba River, and thus the Congo. The volcanic dam also caused a river course reversal in northern Rwanda, generating lakes and marshes.

Climate. Central Africa is characterized by hot and wet climates, which are disposed in a symmetrical manner on both sides of the equator. The equatorial strip that extends from the Atlantic Ocean to the Western Rift Valley is influenced throughout the year by the intertropical convergence zone (ITCZ), resulting in intense, heavy precipitation. The mean annual temperatures in the region vary from 77° to 82° F (25° to 28° C), and monthly means vary only a few degrees throughout the year. The diurnal range of about 20° F (11° C) always exceeds the range of monthly means. Annual rainfall exceeds 80 inches (2,000 millimetres) in coastal Gabon, in the centre of the Congo basin, and on the mountain summits bordering the Western Rift Valley. There is no serious dry season, but in the extreme east and west there are two months (July and August) of lower rainfall.

Also, north and south of the equatorial strip the dry season increases in severity with latitude. It is centred in the Northern Hemisphere in January (when the area receives dry Saharan air masses) and in the Southern Hemisphere in July; the season lasts nearly seven months in the far northern part of the Central African Republic and in far southern Zaire. Notwithstanding its low latitude and proximity to humid air masses, the coastal region of southwestern central Africa has an abnormally long dry season with a low annual rainfall; at Banana, Zaire, at the mouth of the Congo River, it lasts five months. This abnormality is due to the presence of a cold marine current (the Benguela Current) along the coast, which stabilizes the lower strata of the atmosphere and curtails the development of rain clouds but allows the formation of mists. The current reaches as far north as Cape Lopez during the southern winter.

Effect of the Benguela Current

North of the equator, rains are mainly of Atlantic origin. In the south, rains are linked to humid air masses penetrating from the Atlantic Ocean (those not cooled by the Benguela Current during the southern summer and thus holding little moisture), to the humid southern trade winds penetrating from the Indian Ocean, or to the front between these two systems. Throughout central Africa during the rainy season, squalls move from east to west and release "tornadoes" (violent stormy downpours).

Orographic factors greatly modify both rainfall amounts and temperatures; highland areas tend to be wetter and cooler. The slopes and elevated topography, for example, that surround the Western Rift Valley are rainy in comparison with the relatively dry valley bottom (which receives less than 40 inches of rain per year). The western part of Rwanda and Burundi is wetter than the eastern part (*e.g.,* Akagera plain in northeastern Rwanda), which is situated at a lower elevation.

Plant and animal life. Vegetation roughly corresponds to the regional climate, relief, and geology. Equatorial central Africa is covered by an evergreen forest of almost 400,000 square miles. This rain forest—an exuberant world of high trees, rich in epiphytes and lianas—has three main forms: permanently wet marshy forests at the confluence of the Ubangi and Congo rivers; gallery forests, which are subject to periodic flooding, along banks and river floodplains; and, most extensive, forests of dry land, either featuring a single dominant species or, more often, a variety of species (sometimes several hundred per acre). This last type of forest is also found on the eastern slopes of the Congo basin, but it changes to high-altitude lifeforms on the highlands of the Western Rift Valley.

The rain forest is surrounded by a patchwork of savannas and other forests. Savannas of poor, permeable sandy soils unevenly strewn with small trees cover the plateau surfaces, whereas subequatorial forests fill the valleys. Large areas of the savanna probably were created by slash-and-burn agricultural techniques, through which the original forest was cleared and the resulting grasslands maintained by periodic bushfires. Savannas also border the cooler and drier Atlantic coast of Congo and Zaire, while mangrove thickets stretch along the banks of nearby estuaries, lagoons, and deltas.

Beyond this savanna forest region the most distinctive vegetal formation is the dry tropical forest (called miombo in the southeast). Its trees are smaller and less dense than in the equatorial forest, and they are deciduous, losing their leaves during the dry season. The dry tropical forest covers the southern Kwango and Shaba plateaus in Zaire but exists only in shreds north of the equator, in the Central African Republic. More often than not it has made room for wooded savannas. In the far north the thorny steppe is a typical Sahelian landscape.

With its steep gradients over short distances, rainy highlands, and relatively dry bottoms, the Western Rift Valley region is characterized by sharp vegetation contrasts. Central Africa's most famous national park, Zaire's Virunga National Park (formerly Albert Park), is located there. Among its animals are elephants, lions, hippopotamuses, warthogs, forest hogs, okapis, and mountain gorillas on the volcano slopes. Also in the same region is the Kahuzi-Biega National Park, which features mountain gorillas. Other parks are located in savanna or wooded savanna regions: for example, Garamba National Park in northeastern Zaire, which is well known for its white rhinoceroses,

National parks

Upemba National Park in southern Zaire, parks in the northern part of the Central African Republic, and Akagera National Park in northeastern Rwanda. Other parks and reserves are also located in the equatorial forest in Zaire, such as Maiko and Salonga national parks.

Settlement patterns. The most uniformly rainy part of tropical Africa is the least densely populated. In the late 20th century, with the exception of Burundi and Rwanda, the population density in each state was low: the number of persons per square mile was about 10 in Gabon and the Central African Republic, 15 in Congo, and 36 in Zaire. The population is not uniformly distributed, however. Immense areas of the equatorial forest are nearly empty. The obstacle that the forest poses to the African agriculturalist does not explain this situation, for large sections of the more easily cultivated savannas in the north, south, and southeast are also sparsely populated. More heavily inhabited cores are located on the northern fringe of the great forest and at some distance outside its southern fringe. The most densely inhabited region includes the interior slope of the Western Rift Valley and the Burundi and Rwanda highlands, where the density reaches several hundred persons per square mile.

A major transformation of the region has been the development of cities, which as late as the 19th century were limited to a few European coastal settlements. Urbanization was at first slow, but it gathered momentum at the end of the colonial era, especially immediately after independence, which provoked large-scale internal migrations from the countryside to the growing cities.

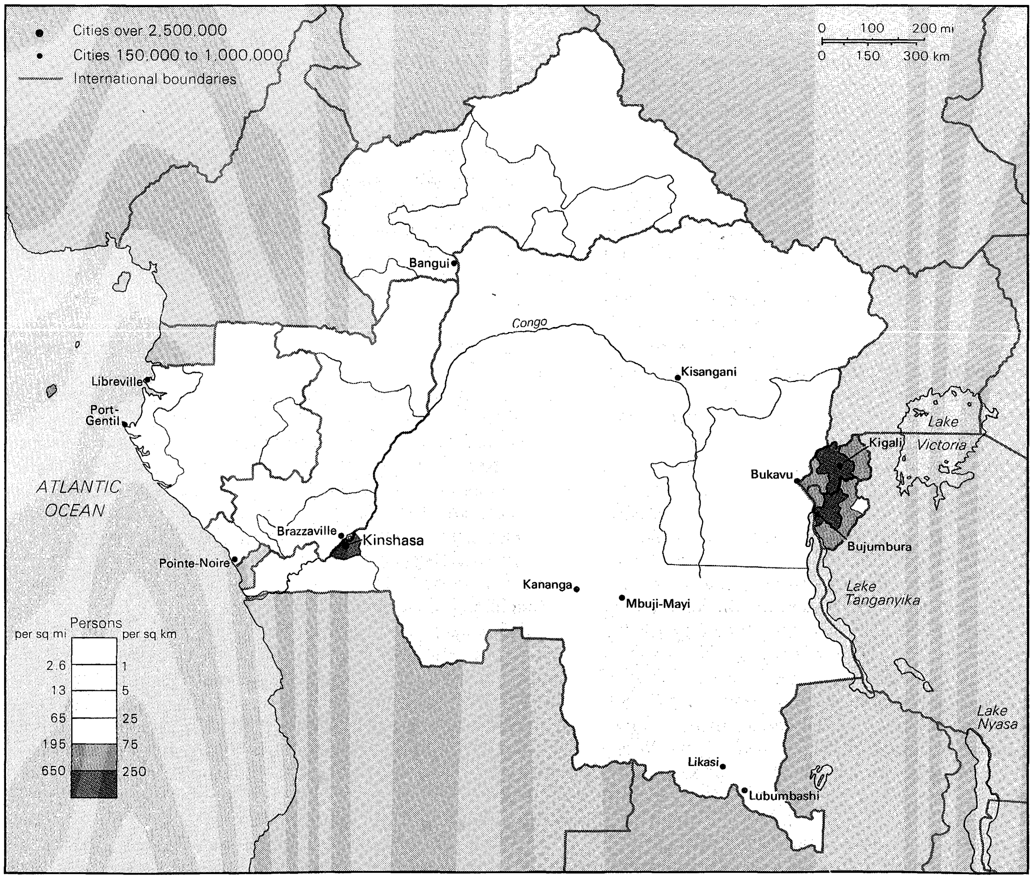
In the late 20th century about 40 percent of the population was urban in most central African countries, with heavy concentrations in the national capitals, provincial administrative centres (*e.g.,* Kananga, Zaire), mining cities (Lubumbashi, Kolwezi, and Mbuji-Mayi, Zaire), and major ports (Port-Gentil, Gabon; Pointe-Noire, Congo; and Matadi, Zaire). Paradoxically, Rwanda and Burundi—the most densely populated countries—have urbanization ratios (6 and 8 percent, respectively) that are among the lowest in the world.

THE PEOPLE

Ethnic composition. National borders have split the territory of many ethnic groups. Pygmies are scattered in the forests from Cameroon to the Kivu and Rwanda mountains. The Fang of Gabon also occupy Equatorial Guinea and southern Cameroon. The Teke are spread throughout Congo, Gabon, and Zaire. The Kongo inhabit western Zaire, western Congo, and Angola; the Chokwe (Tshokwe) and the Lunda occupy Zaire and Angola. In each country some major groups enjoy a numerically dominant position—for example, the Fang in Gabon and the Mboshi, Teke, and Kongo in Congo. Burundi and Rwanda comprise a Hutu majority (agriculturists), a Tutsi minority (pastoralists), and some Twa (Pygmies). In Zaire the major groups are the Kongo (southwest), Mongo (central basin), Luba (south-central), Zande and Mangbetu (northeast), and Ngbaka (northwest). In the Central African Republic, the Baya (Gbaya), in the west, are the most numerous, but the Banda, in the centre, occupy the largest territory. The islands of the Gulf of Guinea, owing to their history of slave plantations and, later, of plantations with a more or less enforced recruitment, have a mixed population originating from Angola as well as from western and central Africa.

Vehicular languages

Linguistic composition. Most languages spoken in central Africa belong to the Bantu group of the Niger-Congo language family. In northern central Africa, Adamawa-Eastern and Sudanic languages are also spoken. The area's rich linguistic diversity includes the use of vehicular languages (*i.e.,* lingua francas) born of local languages, pidgin-creoles (in the islands of the Gulf of Guinea), and European languages, which are the official languages of the various countries (Portuguese in São Tomé and Príncipe, French elsewhere). The vehicular languages often overlap the borders. Lingala, which is important in Zaire, is also spoken in Congo, as are Kongo and Sango, which, in turn, is the main language in the Central African Republic. Swahili, the eastern Africa language, is the lingua franca in eastern Zaire and among the merchants of Burundi and Rwanda. In the latter countries, Rundi (more prop-



Population density of central Africa.

erly, Kirundi) and Rwanda (Kinyarwanda) are (along with French) the official languages.

Demographic trends. As a whole, central Africa has undergone extensive population growth in the latter half of the 20th century. During the 1980s annual growth rates equaled or exceeded 2.5 percent. This rapid growth is a relatively recent historical phenomenon, however. At the turn of the century great epidemics (especially of sleeping sickness)—perhaps worsened by the colonial occupation, which ended ancient isolation—added their effects to those of the other major tropical endemic diseases (malaria, leprosy, and various parasitoses such as filariasis and bilharzia). The population then diminished tangibly. After the 1930s vigorous medical action and a noticeable improvement in hygiene stopped this process and even enabled rapid population growth—except for groups such as the Fang of Gabon, the Nzakara and Zande of the Central African Republic and Zaire, and the Mongo of the Congo basin, whose fertility rates were abnormally low, but which have recently approached those of other groups.

Traditional societies. The Pygmies are hunter-gatherers. Physical characteristics (small stature, thinner lips, clearer skin) distinguish them from their Bantu neighbours, whose language they generally speak. The Pygmies live in symbiosis with their neighbours, supplying them with meat in exchange for agricultural products and iron weapons. They move their campsites frequently throughout the forest regions.

The major part of central Africa's population, however, lives in rural societies with a strong traditional character. These societies are usually organized on a kinship basis of either patrilineal or matrilineal consanguinity (in this case the head of the family is often the maternal uncle). The group that receives a bride must pay compensation in the form of matrimonial possessions to the group which supplied her. Polygyny, in varying degrees, is common almost everywhere.

Land is in principle the collective property of the clan. The headman of the lineage (branch of the clan) is the ruler of the land and divides its use. The rural population most often lives in villages with houses built of local vegetal materials and mud. In contrast to village life, in eastern Zaire, Burundi, and Rwanda the population is dispersed. Each family dwells in an enclosure in the middle of its fields and banana trees.

Religious beliefs

The majority of these populations believe in a creator-god associated with nature spirits. Ancestors, especially the next of kin, are worshiped. A large part of the population is, in principle, Christian, but in many regions the number is lower than 30 percent. Roman Catholicism dominates but is closely followed by Protestantism. The Roman Catholic church is the strongest in Rwanda. Local African Christian churches with a messianic character are of great importance in Congo (Matsouanists, Mpadists) and Zaire (Kimbanguism). Islām has some devotees in the eastern regions—where it was spread during the 19th century by Swahili traders from the Indian Ocean coast—as well as in the north of the Central African Republic. Small Muslim merchant groups also live in the cities.

Traditional political life does not reach above the village level in most forest-dwelling groups. In the savanna, however, there are chiefdoms, varying in size and often with sacral power. Some of these have brought together thousands of individuals, as, for example, the former Kongo, Lunda, Luba, and Kuba kingdoms. Some groups in eastern Zaire, the Rwandese, and the Burundese were organized into kingdoms under a king or mwami. The Zande and the Mangbetu of northeastern Zaire and the Central African Republic built their kingdoms by conquest.

THE ECONOMY

Agriculture, forestry, and fishing. *Traditional agriculture.* Farming societies practice a subsistence agriculture intended to ensure family nutrition. Its rhythm is linked to the rain cycle. Fields are cleared or reclaimed by cutting and burning the forest or the savanna during the dry season and are left fallow after one or more crops. Tools are rudimentary, and there is little, if any, use of manure, which a small stock-farming operation—limited to a few goats or sheep and some chickens (rarely a pig)—cannot in any case supply. No other equipment is used for growing, processing, or transporting the produce. The tasks are distributed between men and women, but the women's participation is substantially greater than the men's. The wife carries nearly the whole burden of the family nutritional subsistence, though the husband participates a little more actively in the trading cultures. Yields are low owing to the general poverty of soils, the absence of fertilizer, and the poor quality of seeds. Heavily populated regions (*e.g.,* parts of Zaire, Rwanda, and Burundi) use more intensive forms of agriculture.

Plantains are the staple food crop in the equatorial forest region; yam and taro are of secondary importance. Since the 19th century, cassava has become more important. Fats are furnished by oil palm trees that often grow wild in the bush. In the savanna regions, in the north as well as in the south, cereal plants such as sorghum and millet dominated in the past, but they have been partially replaced by cassava and corn (maize). Peanuts (groundnuts) are also grown. In the north, sesame is traditionally included in crop rotation. In the highlands (eastern Zaire, Rwanda, Burundi), sorghum, traditional cereals, bananas (used to brew beer), kidney beans, sweet potatoes, and corn are grown.

Stock raising

There is traditionally no cattle raising in most of central Africa, either because this technique was never known by its inhabitants or because of the widespread presence of the tsetse fly, carrier of trypanosomiasis, a disease that attacks both humans (sleeping sickness) and domestic animals (nagana), especially cattle and horses. Stock farming exists in the east and northeast, mainly in the highlands of eastern Zaire, Burundi, and Rwanda, where the tsetse fly is not found. Bororo nomadic stockbreeders, originally from Cameroon, began migrating to the Central African Republic in the 1920s. Pastoralist groups also live in the eastern and northeastern part of this country, into which herds from The Sudan frequently move.

Fishing, an important source of animal protein, is practiced in the interior rivers and lakes and supplies an important trade. There is a small amount of sea fishing in the cool coastal waters.

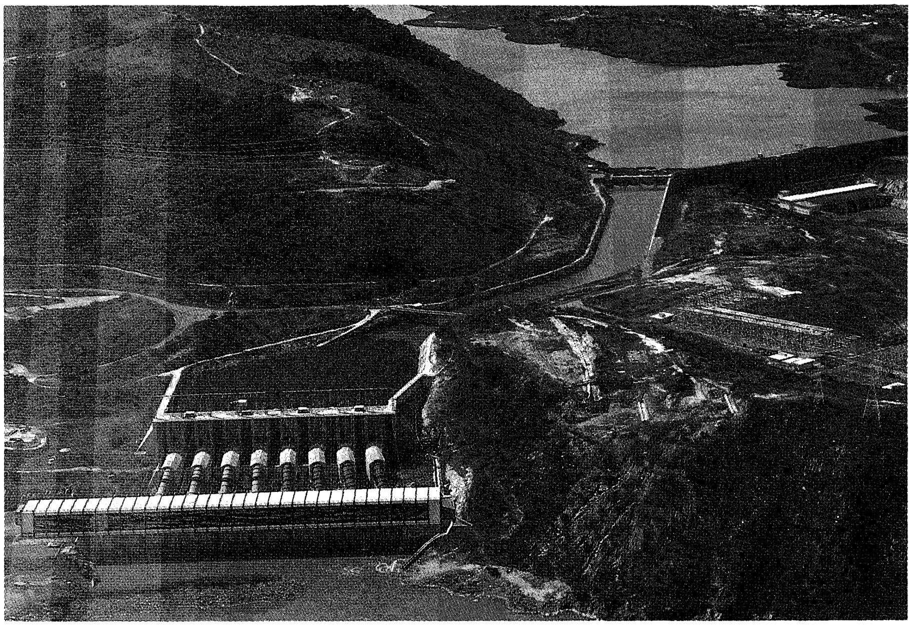
Commercial agriculture. Since the colonial period, agriculture has been part of a monetary economy. From that time, a part of the vegetable production has been intended for the urban markets. Cash crops were often introduced by the colonial authorities: cotton, urena fibres, tobacco, peanuts, robusta coffee in the low regions and arabica coffee in high-altitude regions, cacao, and hevea.

During the colonial era large oil palm, sugarcane, coffee, cacao, and hevea plantations and smaller estates were established by European companies and settlers. Timber was cut in the coastal region forests (especially the limba tree in Zaire and the okoume tree in Gabon) and in the internal forests of northern Congo, southern Central African Republic, and the Lake Mai-Ndombe region in Zaire. Wood was the most important export for Gabon and Congo until the start of petroleum exploitation. Stock-farming ranches were established to supply the cities and, especially, the mining camps.

Agro-industrial models

Most of these estates were returned to nationals after 1960. Several governments have since founded (or helped to found) large agro-industrial companies, for example, in Congo, which adopted a socialist political model with general consolidation of small producer holdings together with centralized planning. But some multinational companies have preserved a dominance in the related area of agribusiness. Unlike most of central Africa, the islands of the Gulf of Guinea continue to be dominated by the plantation system, where cacao plantations followed the 19th-century sugarcane plantations.

Industry. *Resources and mining.* The exploitation of mineral resources contributes substantially to the economy of central Africa. In Zaire, ore fields located in the central and southern regions of the country produce copper, cobalt, manganese, zinc, and tin; quantities of tin are also mined in Rwanda. After independence Gabon was transformed into a mining state, with manganese, uranium, and, especially, petroleum forming the basis of its



The hydroelectric dam on the Congo River at Inga Falls, near Matadi, Zaire.
COMSTOCK INC./Georg Gerster

economy. Gabon was the first petroleum-producing country in central Africa; in Congo, the second largest producer in the region, petroleum accounts for about 90 percent of all exports. There is a modest amount of petroleum production in Zaire. Precious minerals are mined in several countries. Gold washing in some of the region's rivers has developed since 1960, and diamonds are mined in Zaire and the Central African Republic.

Zaire began to tap the hydroelectric potential of the Congo River in the early 1970s through an installation at Inga Falls, about 25 miles (40 kilometres) upstream from the port of Matadi. One of the largest hydroelectric dam projects in the world, the Inga Falls complex provides power mainly for Zaire's mineral processing industries. Other hydroelectric sites in Zaire and Congo have also been developed, mainly in the mining regions.

Manufacturing. Industrial activity remains on a small scale. In the mid-1980s, except for Zaire, Gabon, and Congo, the manufacturing sector accounted for less than 20 percent of the gross domestic product. The main industry consists of the processing of mining products. Major processing centres in Zaire are Lubumbashi, Kolwezi, and Likasi. Sugar mills, sawmills, plywood factories, palm-oil works, cotton ginneries, rubber processing plants, and coffee-dehusking factories are scattered around the countryside. Cement plants are located in western Congo and western and southeastern Zaire.

Other industrial activities are concentrated in the capital cities and the ports of Pointe-Noire, Congo, and Libreville and Port-Gentil, Gabon. These locations primarily manufacture consumer goods such as cotton textiles, shoes, beverages (mainly breweries), tobacco products, soap, and some import substitutes. The low overall standard of living, however, has narrowed the market for consumer goods, hampering the growth of such industries.

Rail systems

Transportation. The 19th-century explorers Henry Morton Stanley and Pierre Savorgnan de Brazza both stressed that without a railway it was impossible to take advantage of the resources of the interior Congo basin, which is insulated from the Atlantic Ocean by rapids and waterfalls. The Matadi-Kinshasa railway began operation in 1898, and the Congo-Ocean Railway between Pointe-Noire and Brazzaville began in 1934. The copper belt of southern Zaire was opened by a web of railways leading to South Africa and Mozambique, to Matadi (by a mixed route of railway to Kasai and then waterway to Kinshasa, where it met the Matadi-Kinshasa line), and to the Angolan port of Lobito (via the Benguela Railway). This web is linked with the Chemins de Fer des Grands Lacs leading to Lake Tanganyika. Rail lines were also laid to bypass the Lualaba rapids upstream of Kisangani, Zaire, thus providing a link between the Lualaba and Congo waterways. In the mid-1950s the Comilog railway was built in west-central Congo to transport manganese ore from Moanda, Gabon, to the port of Pointe-Noire for export. It runs 180 miles south from the Congo-Gabon border (where it is linked to Moanda by a 47-mile cable car system) to join the Congo-Ocean line. The Transgabon Railroad has linked the port of Libreville with Franceville in the eastern interior since 1987; it passes through the important manganese-producing fields near Moanda.

In general the railway, road, and domestic waterway networks have greatly declined in quality since the 1960s. This situation has not been sufficiently matched by the increase of domestic air transport. The Central African Republic continues to suffer from its landlocked location. No rapid transport links give it access to coastal ports; the most direct connections are by road through Cameroon and via the Ubangi and Congo waterway to Brazzaville and then by railway to Pointe-Noire. Rwanda and Burundi send their products through eastern Africa, by road through Uganda or northern Tanzania and then by rail to Mombasa, or, from Burundi, to Dar es Salaam.

Finance and trade. Nearly all industry in central Africa—except for some state-owned companies and a few businesses of mixed public and private ownership—is the product of overseas investment. Before 1960 this came mainly from the colonial power. Some states have nationalized or transferred these assets to their own citizens since independence. These measures have promoted the growth of a middle class that is linked with the political and governmental circles. As a result of poor management, some capital assets in Zaire have been returned to the original investors. Because there is little market for consumer goods, most investment is directed toward export commodities.

Exports and imports

The Central African Republic (cotton), Burundi and Rwanda (coffee), and São Tomé and Príncipe (cocoa) primarily export agricultural products, whereas the main exports of Zaire, Congo, and Gabon are mineral products. Since 1960 the share taken by mining products in these countries has risen as farm exports have fallen owing to inadequate transportation, weak administrative and agronomic management, and the setting of price limits on farm produce. The countries of the European Economic Community and the United States are the leading trading partners. Major imports include petroleum (except in Congo and Gabon) and manufactured products.

Regional cooperation. Difficulties encountered since the accession to independence (disruption of transportation systems, adverse effects of political ideologies and domestic political troubles, drop in prices for agricultural products and sudden fluctuation in bullion quotations, exaggerated growth of the civil service, monetary inflation, and management inadequacy) have led most countries (except Rwanda) to run huge debts. These have been increased, moreover, by the launching of major development projects, some yielding insufficient results (projects linked with the Inga hydroelectric facility or the Transgabon Railroad, for example). For most countries this has resulted in intervention by the International Monetary Fund. Attempts at cooperation in the economic sector remain weak and often link countries that were previously dependent on the same colonial power. Since 1976 Gabon, Congo, and the Central African Republic have been members of the Franc Zone and, with Cameroon, Equatorial Guinea, and Chad, of the Customs and Economic Union of Central Africa (UDEAC). Since 1976, Zaire, Burundi, and Rwanda have belonged to the Economic Community of the Great Lakes Countries (CEPGL). These countries, except Zaire, have also joined the Preferential Trade Area (ZEP). All the central African states are members of the Economic Community of the Central African States (CEEAC), founded in 1983. (He.Ni.)

History

Central Africa consists of three parallel belts of tropical terrain that follow the equator from the Atlantic to the mountains and lakes of the East African borderlands. The northern and southern belts are predominantly orchard woodland with areas of savanna grassland that are either man-made or occur naturally where altitude and soil are appropriate. The middle belt has at most times in the Earth's history been thickly forested, but an increased population dedicated to farming and timber exploitation led to extensive deforestation in the 20th century. One unifying feature of the region is the great Congo (Zaire) River, which rises in the far southeast, crosses the savanna and the forest, captures tributaries from the northern savanna, and recrosses the forest to enter the Atlantic in the southwestern corner of the region. The river has been a source of food and a means of communication over hundreds of generations, but it has also provided an environment in which human and animal diseases have endangered life.

Three historical stages

The population of central Africa has evolved in three broad time zones. During the earliest, which covered a million years, early humans sought food and shelter throughout the savanna regions and probably in the forest as well, though the forest may have been much thinner in the great dry phases of Africa's climatic history. In the second phase *Homo sapiens,* modern man and woman, came to the region and absorbed or eclipsed the thinly scattered original inhabitants over a 100,000-year stretch. The third phase covered less than 10,000 years and brought the development of the societies that have become familiar to modern history. These societies arose from a blend of old populations familiar with the ecological environment and new immigrants with fresh skills to impart.

EARLY SOCIETY AND ECONOMY

Early toolmakers. The oldest population is known almost exclusively from the evidence of its tools. Humankind had made a great intellectual step beyond its fellow primates by learning how to fashion and use tools of a regular form for a specific purpose. The most famous of the Paleolithic tools are the Acheulian knives, oddly known to scientists as "hand axes," used to skin animals and cut meat into chunks that could be chewed raw. Those used in central Africa bear an uncanny resemblance to those used in many other parts of the Old World, suggesting that learning was an intercontinental phenomenon, however slow the transfer of technology may have been. Some of the tools used in central Africa bear marks of local specialization and adaptation, but broadly central Africa was integrated into the culture of the Paleolithic period.

The middle phase of central Africa's prehistory saw significant change, but again the changes suggest that the region was linked to development in other parts of the world. The use of fire to roast vegetable foods as well as meats increased the range of diet and probably resulted in greater human health and fertility. The expanding population may have benefited from new skills in communication. The first use of language enabled societies to become more organized and efficient in their command of natural resources. Mobility in search of food was still the norm, but speech allowed the coordination of effort on a scale that hunting animals could not achieve despite their powers of instinct. The use of tools became ever more varied, though the great majority were probably made of wood or vegetable and animal fibre and so have not survived in the archaeological record. The stone weapons, on the other hand, have survived and show a growing inventiveness among the scattered little bands of central African peoples who survived for millennia in competition with their much fiercer and stronger animal neighbours.

The agricultural revolution. About 10,000 years ago central Africa began to undergo an economic revolution. It started in the north, where a new dry phase in the world's history forced people to make better use of a more limited part of their environment as the desert spread southward once more. Hunters who had roamed the savanna settled beside the rivers and perfected their skills as fishermen. Gatherers who had harvested wild grain on the plains settled beside lakes, where they could sow some of their gleanings as seed in the moist and fertile soils left by the waters that withdrew at the end of each wet season. The northern border of central Africa became one of the cradles of the world's food-producing revolution.

Appearance of vegeculture

The first features of the new way of life in northern central Africa were vegeculture and agriculture. Vegeculture enabled people to collect wild plants on a more systematic basis and to protect the regions where wild tubers grew most plentifully. The regular harvesting of wild roots led to the perfection of specialized digging tools. Stone hoes were ground to a finer polish than the chipped cutting tools of an earlier age. Gradually women and men learned how to clear plots of fertile land and deliberately plant a piece of each root or tuber they ate to allow it to regenerate. They began to select the plant types that most readily lent themselves to domestication, to the ennoblement of regular crops, and to the development of agriculture. The white Guinea yam, *Dioscorea rotundata,* was the basis of the new root farming, which enabled the population to grow in the northern savanna from about 5000 BC.

The second phase of the local agricultural revolution was even more important and had an impact over a wide area of the tropical world. A type of cereal farming based on wild seed of the millet and sorghum families was first developed in the northern savanna. Millet farming became particularly successful in the tropics because, unlike wheat and barley, it did not require the long daylight hours of summer that occur in the temperate climes. Tropical cereals spread from central Africa not only into West Africa but also eastward to India and eventually southward to southern Africa.

The third phase of the food-producing revolution brought an increase in the scale of food production and in its quality. The tending of trees and the gathering of fruit was probably as old as any other form of vegeculture. One of the most valuable of the tree crops was the oil palm, *Elaeis guineensis.* The preparation of palm fruits to make cooking oil enhanced the nutritional quality of the diet with both proteins and vitamins, further enhancing health and leading to population growth and the search for new land to be colonized and cultivated. The tending of trees also enhanced the quality of life in another dimension: some palm trees could be tapped for their sap, and the juice became the basis of a widespread wine industry, adding a festivity to communal life.

A fourth and last aspect of early farming in central Africa was the arrival of a new family of plants. This was the banana family (Musaceae), originally domesticated in the islands of Southeast Asia. Banana plants, like yam tubers, were propagated by cuttings and roots rather than by seeds, but they gradually spread from neighbour to neighbour until the crop had become a dominant one in many parts of central Africa. Banana plants supplied edible roots and textile fibres, but the two fundamental contributions were vegetable bananas (plantains) for cooking and sweet bananas for brewing. The banana flourished particularly well in the wetter areas—in the forest, along the rivers, and in the mountains—and to many societies it became the essential crop. Steamed, baked, fried, or boiled, the banana became the staple carbohydrate of many central African peoples, and they washed it down with a banana beer rich in nutrients.

The agricultural revolution in central Africa was paralleled by another nutritional change as people became more skilled at catching fish. Fishermen—like farmers but unlike hunters—could settle in more permanent village communities. Their diet was richer and more varied. They could own more possessions than simply the weapons and clothes they carried with them. They could make rafts and canoes to transport people and goods on the numerous rivers and lakes. The most important technical innovation was the use of clay for making pots for cooking, brewing, and storing food or drink. A whole "fish-stew revolution" occurred when cooking could be done in earthenware vessels. Pottery also gives the earliest clues about the artistic styles of central Africa; dotted and waved patterns were drawn on pot rims.

The "fish-stew revolution"

The growth of settled communities of hamlets and villages to which food was brought by carriers and canoes led to changes in architecture, social organization, land law and property rights, and warfare and the defense of territory. Although most families probably remained self-sufficient and did their own building, thatching, net making, and pot throwing, some skills may have been concentrated in the hands of a few people of authority or experience. Little evidence survives of Neolithic religion and ritual, but societies surely would have had priests to mediate between the gods and the mortals and to assist with the search for security in an age when the natural forces of fire or flood, plague or famine, were subject only to the most limited human control.

One important change that entered into the Neolithic world of central Africa in the last millennium BC was the spread of new languages. These were spoken in the relatively prosperous and thickly peopled savanna and plateau regions of the Cameroon-Nigeria borderland and came to be known as the Bantu languages, meaning simply the languages of the people, the *bantu*. Languages spread along lines of communication: they followed trade as specialists offered their finest pots or their best axes for sale to their neighbours. They became the vernacular of the clustered villages. They were picked up by countryfolk who came to market. They were carried into new frontier lands by pioneers seeking less crowded spaces. Within 1,000 to 2,000 years the eastern Bantu languages had spread across the northern border of central Africa to reach the highlands of East Africa. Meanwhile, the speakers of the western Bantu languages had found new niches in the great forest and established rapport with the old hunting and foraging communities of both the coast and the river basins.

Development of Bantu languages

Not all the peoples of central Africa were converted to Bantu speech and to the economic way of life of their new neighbours in the first flush of contact. Some hunting societies, commonly known as Pygmies, retained much of their own culture and life-style for another 2,000 years. Some intermarriage occurred with the Bantu speakers of the new villages, and the exchange of hunted meat for grown produce fostered a fairly close symbiotic relationship. Eventually even the Pygmy groups spoke the language of the new majority, which came to outnumber them in nearly all parts of the region. In the deepest forest, however, the hunting societies were able to protect traditional values to a marked degree. Even the great social and political upheavals that accompanied the advent of the Iron Age did not unduly disturb their long-standing and highly specialized relationship with the tropical environment.

The Iron Age. The Iron Age reached central Africa more or less at the same time it reached western Europe, slightly more than 2,000 years ago. The hallmark of the new era was technological innovation, but the social and economic changes that metalworking brought about were fundamental to the agricultural communities—if not to the hunting communities—of the tropics. Ironworking, unlike agriculture, was not a local invention but was a skill that spread from community to community as the superiority of metal tools and weapons came to be recognized. Iron smelting came into central Africa from two directions. In the northwest the oldest source of the new knowledge was on the Nigerian plateau. The skill necessary to dig pit furnaces and surround them with ranks of bellows spread among the Bantu-speaking peoples of the western forest. Gradually the polished stone tools that they had perfected were replaced by much more expensive but effective iron ones. Initially iron was probably used mainly for small and valuable objects such as razors, needles, and knives; later, as the smelting technique became more commonplace, iron came to be used for cutlasses, axes, and, eventually, hoes, which replaced the old wooden digging sticks.

The second source of technical information in central Africa was probably the middle valley of the Nile, where the city of Meroe had been an early industrial site with a huge charcoal industry and great piles of iron slag surrounding its furnaces. The eastern tradition of smelting used furnaces as well as bellows to create the necessary draft with which to turn charcoal and ironstone into wrought iron and molten waste. The iron masters became revered craftsmen and were accorded a quasi-religious status. They lived in some seclusion and often commanded a degree of political authority over their neighbours. Legends of blacksmith-princes became commonplace in the historical folklore of central Africa. Iron became important not only in the immediate locality but also in a developing interregional trade. Although ironstone and charcoal wood were relatively common in most areas, the best smiths could nevertheless command a premium for their wares, and in some regions of deep blown sand or wide alluvial soils, where ores were not available, iron tools and weapons had to be bought from itinerant tinkers.

Iron was a valuable commodity, both raw in wrought bars and worked into spears or machetes. Even more valuable, however, was copper. Unlike northern Africa, central Africa had no Copper or Bronze Age in the last millennium BC, and it was in the Iron Age that the value of copper came to be recognized. Copper was particularly appreciated for its colour and lustre and was used for personal jewelry, rings, bangles, chains, necklaces, and hair ornaments—all made with great craftsmanship and given to persons loved or revered. Copper was also used in the ornamenting of personal belongings, the inlaying of decorations on knife handles, the binding of spear hafts with fine copper wire, and the embellishing of shields with burnished copper nails. Much of the copper mined in central Africa was used to furnish the graves of important people with beautiful objects. The demand for fresh copper rose with each new generation. In particular, the development of new political authorities in the Iron Age led to the need for court regalia for chiefs and kings. Copper was used not only for its visual brilliance but also for the musical quality of the copper instruments that accompanied the nobility on their progress around their domains.

Copper smithing

The three main zones of copper working in central Africa were the Nile watershed, the eastern savanna, and the southwestern forest. Each industrial complex was a thousand miles from the next, and trade intensified as the demand for copper increased. The oldest and largest mines were those of the east. By the end of the first millennium AD, the mines of the Shaba (or "Copper") province of Zaire were casting copper ingots into molds of standard sizes for the international traffic. The region remained one of the world's greatest copper-mining areas for the next thousand years. In the north the copper was used to enhance the wealth and prestige of the local population, but it also fed into the long-distance trade networks of the southern Sahara, particularly after these had been strengthened and organized by Muslim merchants and entrepreneurs in the later Middle Ages. The third copper complex, on the lower Congo River, remained an important but localized industry until a later date. At the end of the 16th century, however, the miners discovered a new outlet for their copper in Europe and sold large quantities to sea merchants from the Netherlands in exchange for Indian textiles, Chinese porcelain, South American tobacco, and stone jars of Dutch gin.

Growth of trade. The development of the copper industry caused many of the peoples of central Africa to look to their own resources for produce that could be sold in order to buy the prestigious new metal and other exotic goods. The salt industry developed out of the needs of long-distance trade. The salt lagoons of the west coast became particularly important, and salt tracks ran far into the interior to agricultural communities without salt of their own to season the cereal dishes which millet growing had made the staple food. Away from the coast, communities that had access to salt mines, or to dried salt lakes, organized exploitation and marketing, sometimes going far afield to the salt mines of East Africa. Where salt was not available, or where the costs of transport made it prohibitively expensive, people had to make do with salt substitutes such as the ash of marsh plants. Where real salt was produced, however, the industry became a focus for political power, and early states were formed, as on the Loango coast of modern Congo.

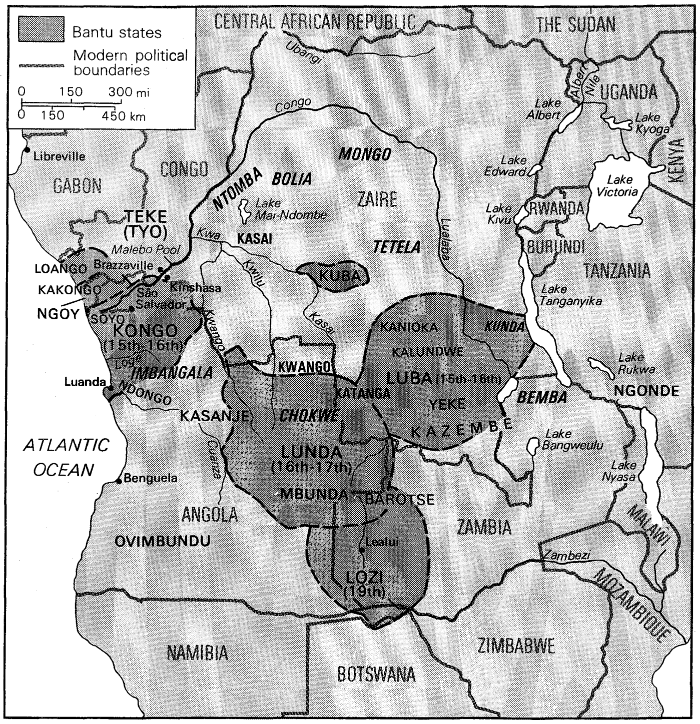
Of comparable importance to copper and salt was the textile industry. This also led to long-distance trade as specialized cloths were made for export to neighbouring regions. Like the salt industry, the textile industry sometimes was controlled by princes who dominated the markets and supplied protection to the caravanners who carried the bales of cloth. In western central Africa the textile industry was based on fibre drawn from the raffia palm. The quality of the finished product, which sometimes had a velvetlike pile and a rich range of natural colours, was much admired by the first foreign visitors to the region. Indeed, raffia squares became the basis of colonial currency in the 17th century. Before that, raffia trading to and from the forest palm groves became one of the sources of wealth among kings in the Kongo kingdom.

Social importance of textiles

Textile weaving in central Africa was important not only to kings and colonial governors but also in regulating social relations. Cloth provided one of the more durable and valuable possessions in every household. It therefore became the preferred item for social payments. In particular, bride wealth—which in other societies might be paid in gold or cattle—was commonly paid in cloth. The control of weaving was in the hands of old men. They no longer had the strength of young men to go hunting or to raid their neighbours, but they did control social wealth by monopolizing the weaving of raffia cloth. Young men had no access to the cloth and thus could not get married and set up a household; they remained dependent on the elders. The old men controlled cloth, and thus marriage, and could in the last years of their lives marry several young wives and ensure that their own line would be dominant in the next generation. In small and scattered societies across central Africa marriage was an important link between communities. Marrying into one's own clan or lineage was not always desirable, and maintaining contacts with neighbours in order to exchange brides was necessary. Permanent overarching political systems were not common, but the regular patterns of social communication went beyond merely commercial ones.

Another source of wealth that became important throughout the history of central Africa was the trade in dried fish. The management of fish ponds became one way in which the scale of political power increased from village size to state size. The lakes of the eastern savanna provide one example of early state formation. The medieval ancestors of the Luba became wealthy and powerful by controlling the fishing industry, building canoes and drying ovens, and setting up networks of trade paths along which porters carried tightly packed headloads of dried fish. In tropical conditions many foods were perishable, but dried fish could be preserved for months and carried to regions deficient in protein, where it was sold for high prices. The fish trade of the great rivers remained important, and fishermen were among the wealthier of the river dwellers. In the 20th century they were able to motorize their canoes, refrigerate their catches, and use steamers for long-distance transport.

In most of central Africa the scale of political management remained small; the power brokers were men who gained influence in local communities. Kingdoms were not the norm. In a few places, however, administration was organized on a scale that went beyond a single day's march from the central homestead. The most striking of the early kingdoms was that of the Kuba peoples, at the heart of the southern forest. Kuba kings controlled the resources of a diverse ecological environment. They built up trade networks that enabled them to obtain copper from far afield and, later, even to buy such valuable ornamental assets as cowrie shells from the Indian Ocean, which were traded from community to community across Africa. The kings of the Kuba based their claim to power on a mystical exotic origin, far away down the rivers to the west, though such an alien pedigree bore no historical justification. Power had more to do with membership in influential local groups than with migration. Wealth, the ownership of livestock, the control of land, the mobilization of militias—these were the features of state formation in those rare areas where kingship could flourish. The Kuba knew government on a scale comparable to the early fish-trading states of the Luba and the raffia-weaving principalities of the Kongo kings.



Bantu states of central Africa from the 15th to the 19th century.

From J. Fage, *An Atlas of African History;* Edward Arnold (Publishers) Ltd.

CENTRAL AFRICA AND THE OUTER WORLD

Development of the slave trade. In the 15th century central Africa came into regular contact with the outer world for the first time. Hitherto all external contact had been indirect and slow. Language, technology, and precious objects had spread to affect peoples' lives, but no regular contact was maintained. In the 15th century central Africa opened direct relations both with the Mediterranean world of Islām and with the Atlantic world of Christendom. The Islāmic contacts remained limited until the 19th century, though the great Spanish Muslim Leo Africanus visited the northern states of central Africa in the early 16th century and described them in Latin for the benefit of the Vatican, where he worked for a time.

The Atlantic opening had an earlier and more direct impact on central Africa than the Mediterranean opening. In the 1470s a colony of Portuguese was settled on the offshore island of São Tomé. The Portuguese had been experimenting with colonial plantations for more than a century and already had settlements on Cape Verde and the Canary and Madeira islands. On São Tomé they es-

Plantations on São Tomé

tablished fields of sugarcane and built sugar mills. This prototype industry, which was later taken to Brazil and the Caribbean, became the richest branch of Europe's colonial enterprise and had a lasting impact on the history of the African mainland. Settlers were unable to build plantations unaided and so recruited local support. European immigrants—predominantly if not exclusively male—sought out African consorts from the adjacent communities and established creole families of plantation owners and managers. They also bought mainland slaves to work the estates. São Tomé became the first bridgehead for the great Atlantic slave trade, which was to have a deep and scarring influence on most of central Africa.

Central African slaves brought to the island slave market were sold to three destinations. The strongest were sold to the Akan miners of the Gold Coast in West Africa, where royal Portuguese agents were able to buy up to half a ton of gold a year in exchange for imported commodities and slave workers from other parts of the continent. A second category of central African slave was shipped to Europe and used both for domestic service in the town and for farm labour on the sparsely peopled estates that Portuguese Christians had conquered from Portuguese Muslims in the late Middle Ages. The third class of slave was put to work locally on the island.

The fortunes of the São Tomé plantations fluctuated over the next five centuries. Sugar gave way to coffee as the mainstay, and coffee in turn was replaced by cocoa in the 19th century. In the 20th century the island was at the centre of a humanitarian storm over the continued use of slaves on the plantations. Cocoa manufacturers boycotted the island, and the planters tried to improve the working conditions of their employees. In the meantime, however, central Africa's premier colony had been eclipsed by other European ventures on the mainland.

The second attempt to build a European colony in central Africa occurred in the medieval kingdom of Kongo surrounding the mouth of the Congo River. Portuguese traders exploited a division in the ruling class to gain a foothold at the court and the support of a royal claimant, who adopted Christianity and assumed the title of Afonso I. The Portuguese had hoped to find precious metals, as they had done in West Africa and were later to do in southeastern Africa, but the only source of profit they could realize was the buying of slaves for the São Tomé market. The king was under increasing pressure to use his army to raid his neighbours for captives. Even the Roman Catholic priests attached to the colonial mission found that they had to finance their activities by trading in slaves. The increasing profitability of slaving, and the lack of alternative sources of exportable wealth, placed growing pressures on the kingdom. Eventually, after the death of Afonso, popular rebellion broke out on a virulent scale. In order to preserve their foothold the Portuguese equipped the governor of São Tomé with an army of 600 musketeers with which to reconquer their mainland base and install a pliant king in the Kongo capital. The Latin American tradition of the Iberian conquistador was thus introduced into central Africa.

The great slaving campaigns of the conquistadores began in the 1570s after the Kongo wars had been quelled. The Portuguese harbour of Luanda was taken over by the Spanish Habsburgs in 1580, when the two crowns were united, and a series of armed assaults was launched on the states of the Mbundu peoples of the interior. The basis of the invasion was the rising demand for slaves to colonize the huge but sparsely peopled provinces of Brazil. Many slaves were captured direct or obtained as ransom for important chiefs. Many more were obtained from long-distance trade networks that penetrated ever deeper into the heart of central Africa both by river and by footpath. The primary imports of the traders were textiles from India, England, North Africa, or Portugal itself. But Portugal, even with the backing of Spain, was not economically strong enough to maintain its monopoly over the foreign trade in central Africa. By the end of the 16th century competitors were frequenting the coast.

Growth of Dutch power

The Dutch were the second colonial power to influence the history of central Africa. Their impact was felt in ways rather different from that of the Portuguese. They were more interested in commodities than in slaves and so opened up the market for ivory. The old hunting skills gained a new value as the market for tusks blossomed and commercial entrepreneurs organized caravans over long distances. Even the shiest of the Pygmy forest dwellers were able to benefit from the hunting of elephants. By the mid-17th century, however, the Dutch had established their own American colonies and so joined in the scramble for slaves. They began to sell guns to their trading partners to facilitate the destruction of old communities and the capture of fugitives. The supply and sale of powder, lead, and muskets became profitable to the coastal brokers but devastating to the inland victims.

The geographic scale of the central African slave trade was enormous. By the 18th century, the supply routes to the Atlantic reached the middle of the continent and began to intersect with the long-distance trade to the Indian Ocean. As the trade spread so did the search for political systems that could manage the traffic. The largest and most successful of the new merchant empires was the empire of the Lunda at the heart of the southern savanna. The Lunda people seem to have become aware of the slave trade as early as the 16th century. Wandering Lunda hunters and salt prospectors, known as Imbangala (or Jaga), entered Angola and recruited local followers into heavily armed bands that raided the countryside, sold their captives to European sailors, and eventually formed an alliance with the Portuguese conquistadores, who allowed them to set up their own kingdom in the Kasanje plain on the borders between Lunda and the European coastal enclaves. At first the kingdom of Kasanje acted solely as merchant brokers to the Portuguese, but with the rise of rival European buyers on the northern Congo coast its network spread farther afield. As the market expanded so did the sources of supply. In the Lunda hinterland a powerful ruler adopting the title of Mwata Yamvo became chief supplier to the Kasanje intermediaries. The Lunda empire spread its commercial network not only to the west but also eastward until it had outlets to the lower Zambezi and the Indian Ocean. The Mwata Yamvo of the west and his viceroy, the Mwata Kazembe of the east, effectively monopolized the slave trade of the heartland. As the Atlantic market grew, Lunda influence spread both north toward the forest and south into the dry plain of the upper Zambezi. In the early 19th century the court began to receive ambassadorial visits from representatives of the king of Portugal, and some years later echoes of the Mwata's greatness reached the Protestant explorer-missionaries of the far south, such as David Livingstone.

In the 18th century the Dutch were replaced by the French as the leading slave merchants on the north coast of the Congo region. The scale of the trade grew rapidly. Forced Congo migrants became the dominant population in Saint Domingue, later called Haiti, which rose to be the richest of all the world's colonies and the largest supplier of sugar. The slaves were carried in crowded, disease-ridden sailing vessels, packed in 500 at a time. Those that survived weeks of suffocation, thirst, and sparse rations on the Middle Passage were sold to rich owners and put to work under cruel bailiffs. The slaves brought some of their cultural values and tried to reconstruct their communities under the shadow of the great plantation houses. Bantu vocabulary and personal names were added to the Creole speech of the Caribbean. Kongo religious practices were preserved in a nominally Christian colonial society when attempts were made to minimize the insecurity and suffering by worship and ritual. But however much the forced migrants tried to reconstruct central African society in the New World, their departure left a serious mark on the old world. The scale of the French trade rose to about 10,000 men, women, and children each year. The demographic hemorrhage was felt in spreading ripples, and the already frail population of central Africa was further weakened. Not until the outbreak of revolution in France, and later in Haiti, did the French trade begin to decline.

Long-term consequences of the slave trade

The long-term consequences of central Africa's involuntary participation in the slave trade affected more than emigration, warfare, and political organization. It also affected agriculture, the occupation of the large majority of the population. In particular, two new crops introduced to the region from the Americas enhanced the productive capacity of the land. The first was corn (maize), which required the same agricultural skills as millet and so could be easily adopted. Corn had the advantage over millet that its grain was wrapped in an envelope of leaves that protected the crop from predatory birds. Corn also had higher yields than millet where soil and water were sufficient, which increased food production and partly compensated for the loss of field hands to the slavers. The second new crop that helped central Africans to recover from the ravages of slaving was cassava, or manioc, a root crop easily adopted by tuber farmers but more difficult for grain farmers to accept. It, too, was better protected from rodents—and even from marauders—than traditional crops. Cassava could be left in the ground when farmers had to flee in war and then recovered whole and edible on their return. To those unfamiliar with cassava, however, it could be dangerous, because the protective poisons in the plant, which made it inedible to vermin, had to be washed out before it was cooked for human consumption.

New crops brought some small benefit to the region from outside, but new diseases had the opposite effect. The growth of long-distance communication led to the spread of smallpox along the slave trails. It was a disease much feared both in the villages and on the slave ships. During serious periods of warfare and raiding, populations were weakened by famine and so easily fell prey to measles and pneumonia. Central Africa was also a malarial zone, and not until late in the 19th century were even the most informed and affluent of traders aware that quinine could protect them. Equally dangerous was sleeping sickness, which periodically spread through the region in great pandemics. As though this were not enough, central Africa was attacked in the early 20th century by the world influenza pandemic. The already weakened population became thinner still and did not recover until well into the middle of the century.

The 19th century did not see the end of the central African slave trade. The Portuguese moved back into northern central Africa when the French trade declined, and Brazil bought more slaves than ever from central Africa in the first half of the century, continuing to use slaves until the 1880s. Spain also came into the market to buy slaves for its surviving American colony in Cuba, where the tobacco industry combined the modernization of the railway age with the old plantation use of slaves. Some of the last slaves also found their way illicitly to the United States, where slaves continued to be employed long after it was illegal to buy and carry them from Africa. Moral condemnation, and the arguments in favour of wage incentives for a more effective labour force, eventually led to the suppression of slave exports by the early 20th century. The end of four centuries of international slaving might have been expected to lead to a new era of freedom and opportunity in central Africa. It did not happen, and instead a whole new set of foreign forces began to penetrate the area. The first of these came from the long-delayed growth of international commerce on the northern border of the region.

Exploitation of ivory. In the second half of the 19th century the northern border of central Africa was suddenly opened up to the impact of a virulent new trade in ivory. The rise of Victorian prosperity in both Europe and North America had led to a rapid increase in demand for ivory to make piano keys, billiard balls, knife handles, and ornamental carvings. Traders from Egypt and the old Ottoman Empire of North Africa came across the Sahara and up the Nile to cross into the upper reaches of the Congo basin, where elephants were still plentiful. In so doing, they severely disrupted local societies, kidnapping local peoples as bearers, servants, and concubines in much the same way that Europeans had done on the Atlantic front. The victims of the trading and hunting razzias not only were used in the heavily armed and fortified ivory camps but also were taken away to be sold as exotic slave girls in the harems of Constantinople or as water carriers in the streets of Cairo.

The second mobile frontier to intrude on central Africa in the 19th century was in the east, and eventually it became as disruptive as the northern incursion. The first immigrants were long-distance traders from the Nyamwezi country in Tanzania who came in search of copper. They set up their own trading kingdom under Msiri and developed a large army of followers equipped with lances and bows. Msiri also trained a military elite of 2,000 men, whom he armed with guns bought on the east coast in exchange for ivory. Msiri's kingdom became one of the largest conquest states to be carved out in central Africa. He adopted the administrative methods of the old Lunda kings, whose provinces he captured and whose governors he reappointed as his own agents and consuls. He also gained control of the old empire's eastern slave trade. In this field, however, Msiri had a powerful rival in the Swahili trade community, which had reached central Africa from Zanzibar.

Trading kingdoms

The Swahili traders and their Arab allies were concerned with both the slave trade and the ivory trade. Their slaves were put to work on the spice plantations of Zanzibar or sold as pearl divers and domestic servants in the Arabian and Persian gulfs. The ivory went to the United States to buy calico, which was in great demand in eastern Zaire. One of the traders took the nickname "Americani" because his American calico was so famed. An even better-known Swahili merchant prince was Tippu Tib, who became the effective ruler of the Swahili towns on the upper reaches of the Congo River. His methods of trade were brutal. Villagers were forcibly rounded up into camps, often with great loss of life—as witnessed by Livingstone on his visit—and then ransomed by their relatives, who were sent out on hazardous elephant-trapping expeditions. The ivory trade thus disrupted the east as effectively as it disrupted the north. Worse still, the pattern of exploitation was one that was soon adopted by the first Europeans to enter the region. They also used capture and ransom to extract wealth from their victims. The first European ruler of the Congo, the Belgian king Leopold, appointed Tippu Tib as his governor and gave him command of the east in recognition of his military and commercial achievements rather than in recognition of his civilizing mission.

The great 19th-century scramble for ivory also brought disruption to central Africa from the south in the years immediately preceding the colonial partition. The agents in the south were Chokwe (Cokwe) hunters from Angola. They had been successful collectors of beeswax, and their trade had enabled them to build up armories of guns, which they eventually turned on their neighbours. They penetrated the heartland of the Lunda empire in the 1880s and destroyed the court. Their victims were sold on the Atlantic coast to be the last European-owned slaves on the old plantations of São Tomé. Their ivory went to the Portuguese after the crown had abandoned its restrictive monopoly on tusks and allowed private entrepreneurs to benefit from market forces. But when ivory became scarce and slaves were frowned upon, the Chokwe pioneered a new branch of trade that was to bring even greater horrors to the peoples of central Africa. This was the search for red rubber, the sap of the wild rubber vine that grew throughout the forest and savanna galleries of the Congo basin. As the price of rubber rose with the development of the electrical and motor industries, so the rubber traders penetrated further into the communities of refugees who had sought to escape the disruptions of the last phase of the slave trade. It was the rubber trade that financed the first stage of formal colonial rule in central Africa.

COLONIALISM

Establishment of European colonies. The pioneer colonizer in central Africa was Leopold II, king of the Belgians. The early attempts of his father, Leopold I, to found colonies in remnants of the Spanish empire in the Pacific or America had failed, and he therefore turned his attention to the heart of Africa, which was still little known to European geographers and therefore less heartily coveted than western or southern Africa. He set up his colony as a private humanitarian venture aimed at limiting the devastation of slaving and the liquor trade.

To finance the venture, however, he rented out nation-size fiefs to commercial companies that were licensed to make a profit and pay tax and tribute to the colonizer-king. Companies such as the Anglo-Belgian India Rubber Company, the Antwerp Company, and the king's own Crown Domain took over the extraction of rubber from the Chokwe. Since the profits on rubber were low compared to ivory or slaves, great pressure had to be brought to bear to encourage newly subjected villagers to forsake their agricultural livelihoods and risk their lives in the forest to gather the vine sap. Military force was used, rubber-collecting became compulsory, and defaulters were barbarically punished by having their hands amputated. The rubber regime in the western Congo was no more benign than the ivory regime that Leopold adopted from the Swahili in the east. The civilizing mission of the first colonizer turned sour very quickly.

Leopold's Congo domain

The humanitarian protest against the rule of Leopold was led by traders who had lost access to their former sphere of interest, by missionaries who deplored the savage denial of human rights, and by a British diplomat who believed in political freedom. Roger (later Sir Roger) Casement publicized the atrocities in the Congo to such good effect that in 1908 the Belgian government confiscated the colony from its own king in an attempt to put an end to the misrule of exploitation. Casement himself went on to a diplomatic career that led to the English gallows when he supported the liberation of Ireland in the same way he had advocated the liberation of the peoples of the Congo, but his humanitarianism did not take root in central Africa. However much other nations might condemn Leopold's rule, rival colonizers were also keen to make their colonies profitable and did so by farming out concessions to private enterprise and turning a blind eye to the large-scale use of forced labour in place of slavery.

The most immediate rival to Leopold in creating instant new colonies was Otto von Bismarck, chancellor of the new German Empire. Most of the colonies he created were outside of central Africa, but he did succeed in laying claim to one tiny but richly populated corner on the mountainous border of East Africa. The old kingdoms of Rwanda and Burundi had thrived for centuries. The ruling class grew tall on the milk of its cattle and governed its farming subjects with imperious superiority. In the forest the old inhabitants continued to maintain their hunting values and, where possible, to escape the attentions of their neighbours. In the late 19th century Burundi underwent severe dislocation, with conflicts over the monarchy and rivalry between chiefs and kings. The Germans moved in from Tanganyika and tried to impose order. They also took over the more stable kingdom of Rwanda, though not for long. In the middle of World War I, the Belgians effectively annexed both kingdoms to their Congo colony and gradually incorporated them into a central colonial administration. The League of Nations maintained watch over the Belgian mandate to care for the interests of former German colonial subjects. The responsibility did not lead to interference, and after World War II the United Nations confirmed Belgian authority and allowed it to continue for nearly another 20 years before the two kingdoms were again separated from one another and given internationally recognized independence.

The French in central Africa

Although the German intrusion into central Africa from the east was slight and short-lived, a comparable French intrusion in the west led to the creation of a much bigger and more lasting equatorial empire. It was the work of the explorer-turned-governor Pierre Savorgnan de Brazza. The roots of the French presence consisted of disused slaving beaches on the Congo coast and a haven for freed slaves called Libreville on the Gabon Estuary. Brazza aspired to join these coastal enclaves to the middle stretch of the Congo River, where the colonial capital was named Brazzaville in his honour. He also aspired to claim territory for France as far east as the upper regions of the Nile. Such an enterprise brought France into competition not only with Leopold, on the far bank of the river, but also with Britain, which had laid claim to the lower Nile in Egypt and wanted to protect the headwaters by conquering the upper Nile as well. The French were narrowly defeated in the race to the Nile but nevertheless gained imperial dominion over a huge stretch of northern central Africa, which they called Ubangi-Shari and which later became the diamond-rich Central African Republic.

The problems that France faced in central Africa were not materially different from those faced by Leopold. The territories were huge, thinly peopled, and poorly endowed with resources that could finance colonial administration and make a profit for the colonizing power. Transport was the greatest difficulty. Leopold went so far as to claim that the essence of colonization was the creation of a transport system. Both France and Leopold were handicapped by the rapids on the lower Congo River and so each had, at huge cost in money and men, to build a railway to reach the navigable middle river. Leopold used a Welshman, Henry Morton Stanley, the Breaker of Rocks (Bula Mutari), to mobilize the necessary forced labour to gain access to his territory. The French tried importing Chinese workers, who could be hired in Asia even more cheaply than local labour could be conscripted in Africa.

The French imitated Leopold—and also the British and Portuguese in south-central Africa—by awarding concessions to colonial companies on the condition that they take responsibility for their own administration and infrastructure in return for the right to extract profits from subject peoples and conquered lands. The most notorious of the French colonial entrepreneurs made their money out of timber concessions. Only toward the end of the colonial period and after did French Equatorial Africa discover that it was rich in iron ore, petroleum, and uranium.

Economic organization. The violent phase of central African colonialism, involving the forced extraction of rubber, ivory, and timber, was followed by a more systematic phase of economic organization. One facet was the establishment of formal plantations on which to grow oil palms and rubber trees. These plantations required capital, machinery, and expensive foreign management. As a result, there was little margin left for adequate wages for workers. The recruitment of labour became the duty of the colonial state or its licensed agents. Some workers accepted the incentive of a cash wage to buy material goods or to accumulate the necessary social payments for marriage. Others were driven into the wage sector by the imposition of cash taxes, which could only be met by working for colonial enterprises. But some were recruited on a compulsory basis—not as convicts deserving of punishment but as subjects who needed to be "civilized" by submitting to a work regime imposed by the state.

Where plantations did not develop, the colonial state found a means of extracting wealth from free peasants. They introduced compulsory crops, most notably cotton—"le coton du gouverneur." Forced cotton imposed severe hardships on farmers, who could not grow family food but instead had to clear land to sow cotton seed for the state. When the crop succeeded, they received a small payment. But much of the cotton regime was applied to marginal lands, where it often failed. The risk was borne by the victim, and famine resulted. The planting of cotton led to frequent protests and to harsh repression in central Africa, as it did in German East Africa and Portuguese West Africa.

The largest industrial complex to develop in central Africa was the mining industry of the copper belt in far southeastern Zaire. Leopold had won a race with the British empire builder Cecil Rhodes to reach the copper mines and had conquered the kingdom of Msiri, killing the king in the process. The next challenge was to build a transport system that could carry machinery into the mining zone and export finished ingots. The politics of transport were the key to development in the high colonial period, as they had been in the early phase. Two railway lines ran to the Indian Ocean coast: one was built across German Africa to the port of Dar es Salaam and another through British Zambesia to the Portuguese port of Beira; a third railroad crossed Angola to the harbour of Lobito on the Atlantic. None of these, however, was under the control of the Belgians, who planned to create a national route to their own port of Matadi on the Congo estuary. To do so, they had to use an expensive and complex mixture of rail and river steamers, and so the copper mines remained integrated in an international network of finance and transport. They also depended on neighbouring regions for their supply of coal and electricity. The industry was dominated by a concession company, the Union Minière du Haut-Katanga, which became almost as powerful as the colonial state itself. The two were interlocked in rivalry and mutual dependence, and the Belgian Congo was described as the "portfolio state" for its reliance on copper shares.

Social consequences of the colonial economy

Two other mining zones added to the wealth of the colonial Congo: diamonds in the west and gold in the east. Between them the three mining zones were large-scale employers of unskilled labour. Some workers were temporary migrants who worked on contract and whose families subsisted on peasants' incomes during their absence. Much of the burden of colonialism fell on women, who became heads of household and managers of family farms when the men were taken to the mines. Part of the mine labour was supplied by migrants who moved permanently to the towns and became proletarian workers exclusively dependent on their wages. Among the townsmen, education and technical training became the road to economic advancement. The Belgian colonial system maintained a rigorous paternal structure, however, and although more subjects in Congo became literate than in other colonies, few could aspire to higher skills or managerial posts. These positions were reserved for the white expatriate population.

One distinctive feature of the colonial era in central Africa was the role of the church. The state provided so few services that missions moved in to make up the deficit. The most famous of all missionaries was Albert Schweitzer, the Alsatian musician and theologian who became a physician and set up a hospital in the heart of French Equatorial Africa. British Baptists also played a major role in converting the people of the lower Congo area to Protestantism and providing them with basic education and minimal welfare services. When turmoil again hit the region at the end of the colonial period, it was the Baptists who brought services to the mass refugee camps. In the Belgian sphere, the Roman Catholic church took a high profile and eventually established a Catholic university through which to train not only colonial whites but also a small elite of black Africans. The rival to the state church was an independent black church, built in honour of the martyred preacher Simon Kimbangu, who spent most of his life in a Belgian colonial prison. Later still, central Africa became a prime zone of evangelism for the fundamentalist groups that sprang up in the United States in the last decades of the 20th century. When the postcolonial state withered, the churches took on greater formal and informal responsibilities for health and education and for communications and financial services in remote areas.

CENTRAL AFRICA SINCE INDEPENDENCE

Continuity and change in politics and economics. The colonial period in central Africa came to an abrupt end in 1960. At a constitutional level dramatic changes occurred. Both France and Belgium decided that they could not resist the winds of change with armed force. Once the black nationalists of West Africa had won the right to self-determination from Britain, it was not deemed possible to deny the same rights in central Africa. New constitutions were therefore agreed, parliaments were elected, and flags were flown and anthems played. General Charles de Gaulle of France, whose path to power had led him to Brazzaville during World War II, became the hero of the new equatorial republics to which he granted independence. King Baudouin of the Belgians visited the Congo for its independence celebrations but managed to orchestrate his reception with less finesse. It soon became clear that "flag independence" in central Africa did not bring any real transformation to satisfy the high aspirations of former colonial subjects.

The continuities of colonial structure after independence could be perceived in a number of different ways. The first president of the former French Congo was a Catholic priest well integrated into French society and culture. Power in the Central African Republic fell into the hands of Jean-Bedel Bokassa (later Bokassa I), a junior army officer who had served the French in a series of colonial wars in defense of the empire; his hero was Napoleon, and his patron was the president of the French Republic. In Gabon power passed to the mining companies, and political influence was orchestrated by officials who worked in the back rooms of the French embassy. When the country tried to change to a nationalist ruler, Paris sent in the French army to restore its chosen neocolonial partner to presidential office. French influence was felt not only through the church, the army, and the corporations, but, most significantly, through the control of the financial system. None of the "independent" French republics had their own currency but were tied instead to the CFA (Communauté Financière Africaine) franc, a former colonial currency managed by France. France also continued to dominate the external trade of the ex-colonies, monopolizing the supply of manufactures and giving itself preferential access to resources, often with a subsidy financed by the European Economic Community.

Instability in Zaire

In the former Belgian Congo (later rechristened Zaire), the colonial continuities were less obvious and the role of Belgium in postcolonial management more volatile. The army, which had been the backbone of order under the police-style administration of the colonial period, reemerged as the single most powerful institution in the country, and its self-appointed leader, General Joseph-Désiré Mobutu (later Mobutu Sese Seko), became the national president after five years of political turmoil in the early 1960s. In the financial sphere Belgium continued to control the copper-mining complex, but the currency was unstable and had to be reformed on several occasions. The Roman Catholic church grew in power after independence and was occasionally seen as a threat to the new elite because of its monolithic and structured capacity to survive in times of social and political decay. In the 1970s many of the expatriate-owned small businesses and plantations were nationalized and given to soldiers, civil servants, and business people, who expected rewards from the state for their loyalty. The result was economic disruption, and expatriates were once again restored to favour. Immigrants from Greece, Portugal, and India continued to play the entrepreneurial role they had exercised in colonial times.

In both the early 1960s and the late '70s the integrity of Zaire was challenged by regional rebellions that threatened the existence of the state. The rebellions were fueled by ideological confrontations over the direction that the postcolonial state should follow. The first rebellion, in Katanga, was backed by investors in Europe and white settlers in southern Africa. They sought to create a mining republic that would not need to share its wealth with the rest of the country and whose black leaders would be amenable to pressure from white neighbours and patrons. Katanga was seen from the south as a buffer that would stop the flow of black nationalist independence movements from reaching British, Portuguese, and Afrikaner territory. The Katanga rebellion failed because the United States decided that a fragmentation of Zaire might give an entrée to the Soviet Union in tropical Africa. Presidents Dwight D. Eisenhower and John F. Kennedy therefore commissioned the United Nations to crush the rebellion by force in a three-year war (1960–63).

A second threat to the survival of Zaire came from the opposite end of the ideological spectrum. Followers of Patrice Lumumba, the murdered prime minister who had gained national independence, and other more radical politicians rose up in a series of rural rebellions. After independence, the breakdown of communications and services led to a serious loss of earnings in the rural areas. Bands of frustrated youths took up arms and attacked the provincial towns. Political leaders who were unwilling to accept the authority of the central government tried to gain help from abroad, in Egypt or eastern Europe. The disorder verged on civil war, and the central government recruited bands of white mercenaries to protect the towns and attempt to restore minimal civil order. In 1964 the United States intervened once again, this time with a direct and controversial parachute drop designed to save

American lives as well as to support the government. By the late 1960s the army government of Mobutu was in control of the whole country.

The third threat to the unity of Zaire occurred in the 1970s. The 1973 revolution in oil prices hit the economy hard since Zaire, unlike Congo and Gabon, produced no petroleum of its own. Once again popular expectations outran economic performance. Worse still, the neighbouring colony of Portuguese Angola threatened to achieve independence under a left-wing government that advocated the economic rights of peasants and workers. Mobutu, probably in collusion with the United States and South Africa, decided to lend part of his army to the conservative party in Angola to ensure that a right-wing regime would be installed, rather than one which might encourage renewed rebellion among the dispossessed of Zaire. The venture failed—in spite of the support of a South African armed column and the U.S. financing of renewed mercenary involvement in central Africa—and Mobutu temporarily lost his credibility as a strongman. Katanga, renamed Shaba, rebelled twice more, and foreign troops were flown in to maintain the authority of the government. This time the initiative was taken by France, and the soldiers came from Morocco in exchange for political concessions in North Africa. Mobutu was again restored to preeminence and went on to celebrate 25 years of power in 1990.

In the 1980s central Africa managed on the whole to avoid the world's limelight. Economies developed at a modest pace. Engineers made increasing use of the hydroelectric power of the Congo River and carried it over greater distances to maintain the industrial complexes inherited from the colonial era. In Gabon a new railway was built to transport logs from one of the world's last great forests and to reach new mining complexes of the far interior. The world price of iron ore dropped, undermining the extractive plans of the various governments.

Role of petroleum in developing economies

The most serious difficulty for central Africa was the decline in the price of petroleum. The price rises of 1973 and 1979 had been hard on petroleum-importing Zaire but had benefited petroleum-exporting Congo and Gabon. The decline in price in the 1980s had the reverse effect. Gabon had become particularly dependent on its oil revenue. The city of Libreville had attracted about half the country's population; the oil boom also brought 60,000 migrants from Equatorial Guinea. So close did relations with the former Spanish colony become that it adopted the CFA currency of Gabon as its own. The intensive urbanization of Gabon was partially matched elsewhere in the region. Nowhere did agricultural production meet the rising demand for foodstuffs in the towns. Cities had to be fed with imported staples paid for with foreign exchange. Gabon saw an influx of petty traders from Nigeria as well as the expansion of the import-export businesses controlled by Lebanese and Senegalese entrepreneurs. The countryside was denuded, and farming fell into abeyance.

One corner of central Africa that did occasionally catch the eye of the media was Burundi. The lack of satisfaction brought by independence was particularly acute in the overpopulated highlands. Demographic pressure led to increased competition over land resources. The conflict took on class and ethnic dimensions, and spasmodic fighting broke out between the aristocratic Tutsi and the farming Hutu. They vied with one another to gain influence on government—the largest employer, the purveyor of educational opportunity, and the source of all power. The relatively fertile and healthy highlands remained overpopulated, and the children remained undernourished. In the 1980s a new plague, already endemic in neighbouring Uganda, added AIDS to the existing catalog of diseases. Throughout central Africa the human immunodeficiency virus (HIV) spread rapidly in a society where traditional family norms had been disrupted and sexual partnerships were often casual or financial. Urbanization grew apace and with it a loss of moral control by village elders, making it impossible to predict what the health record of central Africa was likely to become.

Postindependence culture. One of the most lasting legacies of the colonial period was language and culture. In the French territories an assimilated elite spoke French, married into French society, adopted French styles of clothing and eating, created French universities with French syllabi, and wrote French prose and poetry in praise of their blackness while rejoicing also in their Frenchness. French was the language of high society in Zaire as well, but two other languages that had been used for colonial command survived alongside the local vernaculars. In the east, Swahili had been the language of the military and became a widely used lingua franca. In the west the language of the missionaries, Lingala, served to bridge the communications gap between town and country and between the numerous tiny ethnic groups. But in other ways European culture took root. Musical instruments in particular were adapted to the powerfully evolving local musical tradition. Zaire became the home of the most famous African bands. Although the country could not itself mass-produce records and discs, its musicians became a part of world culture and their music spread out from the great studios of Europe that bought up the recordings. Zairian dance music became one of the best-known and happiest of the exports from central Africa.

Influence of central African art

A second export from central Africa that has greatly enriched the culture of the world is its art. In the late 19th and early 20th centuries several great explorers—including Henrique de Carvalho from Portugal, Oskar Baumann from Austria, Émil Torday from Belgium, and Percy H.G. Powell-Cotton from England—brought back chests of beautiful objects with which to fill enthnographic museums. The plastic arts of central Africa excited particular curiosity, and wooden statues of great originality were bought from the Kuba and the Chokwe and eventually came to be auctioned as objets d'art of great price. The influence of central African art was particularly felt by the Spanish artist Pablo Picasso and European painters of his school. Artistic influence was a two-way process, however, and societies in which decoration and ornamentation were important also adopted European motifs and artistic materials. Pottery designs long remained traditional, but the burned pictures drawn on gourd vessels began to show the material objects of the new age. Art was also a therapeutic means of coming to terms with the new wave of potentially threatening invaders brought by colonialism. In a more peaceful vein, a colonial school of art developed, with distinctive pictures of hunting and fishing scenes painted in brilliant colours on black paper. In central Africa art remained a vital part of modern culture as well as of deep history. (D.Bi.)

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 944, 96/11, and 978, and the *Index.*

THE COUNTRIES OF CENTRAL AFRICA

Burundi

Landlocked, beset by population pressure and meagre economic resources, the Republic of Burundi (French: République du Burundi; Rundi: Republika y'u Burundi) is one of the poorest and most conflict-ridden countries in Africa. Shaped in the form of a broad arrowhead, it is bounded on the north by Rwanda and by the converging boundaries of Tanzania to the east and Zaire and Lake Tanganyika to the west. Its small size (10,747 square miles [27,834 square kilometres]) belies the magnitude of the problems it faces in seeking to reconcile the supremacist claims of the ruling Tutsi minority with the growing demands for political participation of the Hutu majority.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief and drainage.* Topographically, Burundi is the prolongation of a chain of mountains and

Burundi highlands

high plateaus forming the eastern flank of the Western Rift Valley. In the northwest the narrow Imbo valley extends southward from Rwanda to Lake Tanganyika and includes the Rusizi (Ruzizi) River, which separates Burundi from Zaire. Farther south, along the shores of Lake Tanganyika, the land rises steeply to form part of the Congo-Nile divide, which in places reaches altitudes of 8,500 feet (2,600 metres).

Soils. Light, forest-derived soils predominate, forming a thin layer of humus over lateritic subsoils. Because of the steep slopes and frequency of rainfall, soil erosion is a serious problem. The leaching away of the fertile topsoil is a major constraint on agriculture. The best soils are formed from alluvium, but their extent is confined primarily to the lower portions of larger river valleys.

Climate. Altitude is a major factor in Burundi's climate, greatly moderating its tropical character. The country's high average elevation produces relatively cool temperatures, which average only about 70° F (21° C) throughout the year in the central plateau area and usually drop to below 60° F (15° C) at night. At lower elevations the annual average is only slightly higher—73° F (23° C), for example, at Bujumbura in the Imbo valley. Annual rainfall averages 60 to 70 inches (1,500 to 1,800 millimetres) in the highest-lying areas but only 40 inches on the shores of Lake Tanganyika. Except for a short summer dry season, rainfall is evenly distributed throughout the year.

Plant and animal life. The natural forest vegetation has almost entirely disappeared from the landscape and is limited now primarily to higher mountain slopes. On the plateau, wooded savanna is found at higher elevations, giving way to more open savanna on the lower slopes. Poaching has dealt a severe blow to the country's wildlife. The elephant population has virtually disappeared, leaving only waterbuck, warthogs, baboons, and antelope among the less endangered species.

Sarah Errington/The Hutchison Library

Hutu family compound, Burundi.

Dispersed pattern of settlement

Settlement patterns. The standard pattern of settlement is one of family compounds, with characteristic beehive huts rising above the hedges of individual enclosures. Villages are almost nonexistent; dispersal rather than concentration is the key characteristic of settlement patterns. The population density is one of the highest on the continent. Bujumbura, the capital, sprawled along the northern tip of Lake Tanganyika, is the largest city.

The people. *Ethnic composition.* As in Rwanda, Tutsi and Hutu are the principal ethnic communities, representing respectively nearly 15 percent and more than 80 percent of the total population. Other groups include the pygmoid Twa (representing less than 1 percent of the population), the descendants of the princely families (*ganwa*), and a sprinkling of Swahili-speaking peoples from Tanzania and Zaire. Popular images of Tutsi as uniformly tall and graceful, and Hutu as short and stocky, do not fit the reality of physical variations. A more meaningful distinction refers to ethnoregional identities, as between Tutsi-Hima, in the southern province of Bururi, and Tutsi-Banyaruguru, found primarily in the north. The former constitute the bulk of the armed forces.

Linguistic composition. Rundi (Kirundi), the standard medium of communication throughout the country, has joined French as an official language. Swahili, the language of trade, is, along with French, widely spoken in the capital city.

Position of the Roman Catholic church

Religion. The country claims a large Christian population, of whom about 90 percent are Roman Catholic. In recent times church-state relations have become the focal point of renewed ethnic tension. Under the Second Republic (1976–87) a variety of sanctions were taken against the Roman Catholic church in an attempt to curtail its social and educational activities, as these were thought to favour the Hutu against the Tutsi. After the coup d'état of 1987, the issue was temporarily defused, yet the church continues to be seen by many Tutsi as a dangerously subversive institution.

Demographic trends. Although infant and child mortality rates are high, Burundi's birth rate is above average for central Africa, and its population is growing at about the same high rate as Africa as a whole. More than 40 percent of the population is under the age of 15, assuring a continued high growth rate. Less than 8 percent of the population is urban, most of whom live in Bujumbura, the only city with a population of more than 100,000. Life expectancy at 49 years is below average for the continent.

The economy. Agriculture is the lifeblood of the country, with industrial activities accounting for less than 10 percent of the gross national product (GNP). Coffee, chiefly arabica, is the principal export crop and source of foreign exchange. Cash crops of lesser importance include cotton, tobacco, tea, and pyrethrum (a flower used in making some insecticides). Rural poverty is evidenced by the predominance of traditional subsistence agriculture, the persistence of low income levels, chronic deficits in the balance of trade, and heavy dependence on foreign aid.

Resources. Unexploited mineral resources include huge nickel deposits in the eastern region, as well as significant vanadium and phosphate deposits. Geologic assessments also indicate possible major petroleum reserves beneath Lake Tanganyika and in the Rusizi valley. Mineral production, however, is limited to small quantities of gold, tin, and wolfram. Peat and firewood are the two major local sources of fuel.

Coffee cultivation

Agriculture, forestry, and fishing. Approximately 50 percent of Burundi's land area is considered cultivable, and 35 percent is suitable for pasture. Staple food crops include beans and corn (maize), grown in the wet season (September to December), and cassava and sorghum, grown in the drier part of the year (January to July). Arabica coffee is grown by many farmers, reaching an average output of 30,000 to 35,000 metric tons per annum. Large areas of cotton are cultivated, mainly in the Imbo valley. While the density of livestock results in overgrazing, the commercial value of livestock production is virtually nil: on the average a cow yields only 40 gallons (150 litres) of milk per annum, and the dressed weight of cattle is only about 250 pounds (110 kilograms). Despite efforts at reforestation, the country's forested area has shrunk to less than 3 percent of the total area. Fishing activities are conducted on Lake Tanganyika and in the smaller lakes and rivers of the interior, both rich in tilapia. About 13,000 tons of fish are caught annually.

Industry and finance. Industry is limited to small-scale processing and manufacturing plants, concentrated mostly in Bujumbura. The government-owned BRARUDI brewery and COTEBU textile company are the two largest industrial enterprises. BRARUDI produces some 23 million gallons of beer annually, providing the government with about one-third of its tax revenue. With the sharp drop in coffee prices, tax revenue declined drastically in the 1980s, in turn causing huge annual budget deficits.

Trade. The trend in the 1980s was in the direction of a growing trade deficit and heavy dependence on foreign aid. On the average, export earnings are only 50 percent of the import bill, reflecting a steady growth of consumption and investment coupled with a sharp decline in the international price of coffee and rising import prices. Approximately 90 percent of Burundi's export earnings come from coffee, with cotton and tea accounting for most of the remaining exports.

Transportation. In the absence of railroads, only three major routes are available: the northern route by road from Bujumbura to Mombasa, Kenya, via Rwanda; the central route by barge down the Rusizi River to Lake Tanganyika, then to Kigoma, Tanz.; and the southern route across Lake Tanganyika to Kalemi, Zaire. A secondary road network connects Bujumbura to various provincial capitals. In addition to Bujumbura's international airport there are five smaller airstrips for domestic flights.

The Third Republic

Administration and social conditions. *Government.* Since the advent of the Third Republic, on Sept. 3, 1987, Burundi has been ruled by a 30-member military junta, the Military Committee for National Salvation (Comité Militaire de Salut National; CMSN), headed by President Pierre Buyoya, a Tutsi-Hima. Despite the appointment of a Cabinet consisting of an equal number of Hutu and Tutsi, headed by Prime Minister Adrien Sibomana (a Hutu), the CMSN is the key decision-making organ. Pending the adoption of a new constitution, the ruling party, the Unity for National Progress (Unité pour le Progrès National; UPRONA), acts as the principal counterweight to the army. Burundi is divided into *provinces, districts,* and *communes.* Power at the local level rests in the hands of centrally appointed authorities, virtually all of whom are of Tutsi origins.

Education. Primary education begins at age seven and is compulsory for six years; secondary education, divided into programs of four and then three years, is not compulsory. The distribution of the school-age population shows a striking disproportion in enrollment figures between primary and secondary schools, the former accounting for about 90 percent of total enrollments. Only a small fraction of primary school students are admitted to the secondary level, and fewer still are able to gain admission to the University of Burundi, at Bujumbura, the country's only university. Ethnic discrimination in schools remains a politically explosive issue. The overrepresentation of Tutsi at the secondary and university levels translates into the absence of significant avenues of upward mobility for Hutu, resulting ultimately in a virtual monopoly of civil service positions by representatives of the Tutsi minority.

Health and welfare. The most common health problems stem from communicable diseases and nutritional deficiencies, a situation that accounts for 70 percent of infant and child mortality. Malaria, measles, influenza, and diarrhea account for 80 percent of reported cases of communicable diseases. While sleeping sickness is widespread in the lakeshore areas, pulmonary diseases are common in the central highlands.

Cultural life. *The arts.* Much of Burundi's rich cultural heritage, most notably folk songs and dances, was intended to extol the virtues of kingship; since the fall of the monarchy in 1966, however (and particularly after the bloodbath of 1972), cultural life has entered a period of prolonged lethargy. The annual sorghum festival (*umuganuro*) was once the occasion for a magnificent display of traditional dances by court dancers (*intore*), with the Karyenda drum, emblem of the monarchy, intended to give both musical and symbolic resonance to the ceremony.

Cultural institutions. Despite the efforts of the government to promote interethnic harmony through displays of a shared cultural heritage, the latter assumes modest proportions. Arts and crafts, including basketry and beadwork, are prominently displayed, however, in the country's only cultural centre.

Press and broadcasting. Radio and television are under strict government control, and so is the press. The only existing daily, *Le Renouveau du Burundi,* is also the official mouthpiece of the government. Because of the strict censorship exercised by the government, tracts are the primary medium through which members of the opposition (both Hutu and Tutsi) make their views known to the public.

For statistical data on the land and people of Burundi, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Origins of the monarchy

The origins of the Burundi monarchy are shrouded in legend. According to one set of traditions, Ntare Rushatsi, the founder of the original dynasty, came from Rwanda in the 17th century; a more reliable source suggests that Ntare must have come from Buha, in the southeast, and laid the foundation of the original nuclear kingdom in the neighbouring Nkoma region.

Until its overthrow in 1966, the kingship was the linchpin of the political system and the most obvious element of continuity with the past. Yet the type of monarchy that emerged from the debris of the Belgian trusteeship at the time of independence in 1962 had little in common with the weakly structured and highly decentralized archaic kingdom in existence at the time of the German conquest in 1903.

Although Belgian rule did little to mitigate factional competition between the two rival dynasties—the Bezi and the Batare—the legislative elections of 1961 resulted in a landslide victory for the representatives of the ruling Bezi dynasty, identified with the Union and National Progress Party (Parti de l'Union et Progrès National; UPRONA), whose leader at the time was Prince Rwagasore, the eldest son of Mwami (King) Mwambutsa. Rwagasore stood as the embodiment of populist aspirations and the strongest supporter of the monarchy. His assassination on Oct. 13, 1961, ushered in a crisis of legitimacy from which the country has yet to recover.

The turning point came on Oct. 18, 1965, when a group of Hutu officers unsuccessfully tried to overthrow the monarchy, yet came close enough to realizing their objective to cause the panic-stricken Mwambutsa to flee the country. By then the inability of the crown to handle the competing claims of Hutu and Tutsi had become palpably clear. The assassination of the Hutu prime minister Pierre Ngendandumwe, on Jan. 15, 1965, by a Tutsi gunman, provided dramatic evidence of the spillover of ethnic hostilities from the party into the government. The critical factor behind the abortive coup, however, was Mwambutsa's decision to ignore the verdict of the May 1965 elections, which gave the Hutu 23 seats in the National Assembly out of a total of 33. By appointing his private secretary, Léopold Biha, as prime minister, Mwambutsa made it unequivocally clear that power would continue to rest with the crown.

Mwambutsa's abdication

In the complicated sequence of events that followed the abortive coup, the execution of some 34 Hutu officers was the first in a series of steps intended to give Tutsi elements unfettered control over the institutions of the state. The second came in July 1966 with the appointment of Michel Micombero as prime minister. A Tutsi-Hima from Bururi province, Micombero had played a key role in thwarting the 1965 coup and in organizing anti-Hutu pogroms in the countryside. Finally, with the formal overthrow of the monarchy in November 1966—less than three months after the official transfer of the crown from Mwambutsa to his younger son, Prince Charles Ndizeye—and the formal proclamation of the First Republic, the last obstacle in the path of Tutsi hegemony was removed.

No other event has cast greater discredit on the First Republic than the genocidal killings perpetrated against the Hutu community in April and May 1972. The carnage took the lives of approximately 5 percent of the population and virtually wiped out the educated stratum of the Hutu population. Anywhere from 100,000 to 150,000 Hutu were killed in the course of the repression triggered by the outbreak of Hutu-instigated uprisings.

Besides creating deep and lasting hatreds on both sides of the ethnic divide, the events of 1972 became the source of considerable tension within the Tutsi minority, thus paving the way for the overthrow of Micombero in 1976 and the advent of the Second Republic under the presidency of Jean-Baptiste Bagaza. Though himself a Tutsi-Hima from Bururi, like Micombero, Bagaza set out to reinvigorate the UPRONA on a scale that Micombero had never contemplated. At the same time every effort was made to bring the Roman Catholic church firmly under the control of the state.

The crisis in church-state relations was the critical factor behind Major Pierre Buyoya's decision to overthrow the Second Republic on Sept. 3, 1987, and proclaim the advent of the Third Republic. The 1987 coup did more than replace one group of Tutsi oligarchs by another; it signaled an important shift of policy on the issue of church-state relations, and, by implication, on the Hutu-Tutsi problem. Ironically, Buyoya's call for liberalization, while significantly raising the expectations of the Hutu masses, did little to alter the rigidly discriminatory posture of Tutsi civil servants in the provinces. The gap between Hutu expectations and the realities of Tutsi hegemony lies at the root of the killings of August 1988.

Informed estimates put at 20,000 the number of people killed in the northern communes of Ntega and Marangara between Aug. 15 and 25, 1988, the overwhelming majority of Hutu origins. As in 1972 the initial outburst of violence—in the wake of countless provocations by local Tutsi officials—came from Hutu elements. Unlike his predecessor in 1972, however, President Buyoya responded to the crisis in a surprisingly conciliatory fashion. For one thing, the existence of a Hutu-Tutsi problem had been explicitly recognized by the Bujumbura authorities, along with the need for appropriate solutions. Moreover, a conscious effort had been made to achieve parity of ethnic representation within the government. Finally, and most importantly, a National Commission in Charge of Studying the Problem of National Unity was entrusted with the task of making specific recommendations to the government to "protect and strengthen the unity of the people of Burundi."

Search for national unity

Whether as a result of these recommendations significant changes will take place in the distribution of rank and privilege remains unclear. Although the formal institutions of government are no longer a monopoly of the Tutsi minority, the army and the police remain firmly in Tutsi hands. What is beyond doubt is that memories of the 1972 and 1988 slaughters will persist for generations and so will the mutual fears and hatreds they have instilled in the minds of both Hutu and Tutsi. (Re.L.)

For later developments in the history of Burundi, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Central African Republic

The Central African Republic (République Centrafricaine) is an independent state located in the centre of Africa. It has an area of 240,324 square miles (622,436 square kilometres) and is bordered by Chad to the north, The Sudan to the east, Zaire and Congo to the south, and Cameroon to the west. The capital, Bangui, is situated on the southern boundary, formed by the Ubangi (Oubangui) River, a tributary of the Congo (Zaire) River.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief, drainage, and soils.* The Central African Republic occupies a vast rolling plateau that rises to an altitude of between 2,000 and 2,500 feet (610 and 760 metres) above sea level and that forms, along a southwest to northeast trending crest, the drainage divide between the Lake Chad and Congo River basins.

Topography

The vast central plains rise gradually to the Bongos (Mongos) Massif (4,600 feet) to the northeast and to the Tondou Massif in the east and southeast. In the west, the Karre Mountains (4,000 feet) form a high granite range that declines eastward into sandstone plateaus. In the north, the most important mountains are those of the Dar Challa range, which rise to 4,423 feet (1,348 metres) at Mount Tinga on the border with The Sudan. In the southeast is a plain cut by a number of rivers.

Climate. The climate is tropical. The Central African Republic lies in a transitional climatic zone between the sub-Saharan zones to the north and the equatorial zone to the south. During the wet season (from March to October or November) heavy rainstorms occur almost daily, and early morning fog is typical. Maximum annual rainfall occurs during August and September, especially in the upper Ubangi region (71 inches [1,800 millimetres]) and in the Karre Mountains (59 inches). During this season of southwestern monsoon (rain-bearing) winds, the daily temperature ranges between 66° and 86° F (19° and 30° C).

The dry season—brought by the northeastern trade winds, or harmattan—generally begins in October and ends in February or March. The air is dry, and temperatures range between 64° and 104° F (18° and 40° C); it is warm during the day but considerably cooler at night. The skies are clear and seldom cloudy, although some dry fog can be observed. Sandstorms or dust storms also occur in the extreme north.

Plant and animal life. The country lies largely in the savanna zone of Africa. While the north is treeless, dense rain forests follow the southern rivers and stretch far into the savanna country. Bushfires in the dry season are a constant threat to the republic's extensive timberland, which is densest in the far southwestern regions.

Animal life

Antelope, buffalo, and elephants are found on the savanna. Along the rivers, which are abundant with fish, groups of hippopotamuses, crocodiles, and rhinoceroses are to be found. A rich and varied birdlife—in addition to many varieties of snakes, bats, and insects, including many colourful butterflies and moths—makes the territory zoologically one of the most interesting in Africa.

Settlement patterns. More than 60 percent of the total population is rural, living along limited zones of the Ubangi River and along the Chad border. The eastern and northeastern sections of the country are thinly populated. With more than 40 percent of the people under 15 years of age, the rapidly growing population will approach four million by the end of the century. Much of this growing population gravitates toward urban areas, the largest of which is Bangui, the capital city on the Ubangi River. Other major towns are Berbérati, Bossangoa, and Bouar, in the west; Bambari and Bria, in the central plains; and Bangassou and Mobaye, on the Ubangi.

The people. *Ethnic and linguistic composition.* The people of the Central African Republic compose an interesting ethnic mosaic. Ranging from the most simple hunting-and-gathering groups of Aka to state-forming groups such as the Zande (Azande) and Nzakara, most Central Africans had developed a relatively homogeneous culture through centuries of interaction by the time of the European conquest.

Shifting ethnic identities

The attempts by colonial administrators and ethnographers to fix Central Africans into hard and fast ethnic groups no longer seem viable. Central African ethnic groups are and have been historically open. Their membership and identity shift constantly. In this heterogeneous and environmentally diverse area, ethnic groups have come together, changed, and disappeared continuously throughout the past and continue to do so.

Beginning under the French colonial regime, a small bureaucratic elite drawn largely from riverine people, who—as slavers, fishermen, and traders—had served other groups as middlemen, emerged as an indigenous ruling group for the whole country. Since independence this elite group of Ubangi region ethnics, such as the Mbaka and Yakoma, has held most positions of political power. Other Central Africans, far more numerous, have tended to resent this situation and have occasionally rallied to nonriverine leadership, reflecting latent ethnic hostilities.

A handful of Greek, Portuguese, and Yemeni traders are scattered around the country. A small European minority, largely French, living in Bangui, as well as small numbers of more recent political refugees from the southern Sudan and Zaire complete the population of the country.

The national language, Sango, is a lingua franca that developed during the course of the 19th century; it has many dialects. A written form is used in mission publications and in some schools. The official written language is French, which is taught in the schools. Other ethnic languages, such as Baya (Gbaya) and Mandjia, are widely spoken. The oral history tradition is not strong in this often-disrupted country.

Religion. The vast majority of Central Africans are roughly equally divided among putative adherents of Roman Catholicism, various Protestant denominations, and traditional African religious practices. There is a small Muslim minority and an undetermined but large and growing minority of Central Africans with little if any religious affiliation.

Demographic trends. As is the norm in Africa, the Central African Republic is underpopulated. Its inhabitants are increasing at a rather rapid rate, yet, given the country's low population density, relative lack of immigration, and high infant death rate, this growth rate does not seem to pose a problem in the short run. Much of the population is under the age of 15 and is poorly nourished. Life expectancy is only 44 years for men and 47 years for women. Far too much of the available land in the Central African Republic is given over to the cultivation of crops, such as cotton, that are destructive in tropical areas. As a result, a growing number of Central Africans suffer from malnutrition. Cases of kwashiorkor, marasmus, and other malnutrition-induced diseases are becoming the norm in urban and more densely populated areas.

The economy. Agriculture is the basis of the Central African economy and provides the livelihood for most people. A large percentage of the population are subsistence agriculturalists producing cassava (manioc), millet, sorghum, and peanuts (groundnuts) for their own consumption. International (mostly French) capital dominates the economy. The Central African Republic is a classic neocolonial situation: the national economy is controlled by both private and public investment from the former colonial power, and a small minority of local elites live on the proceeds from international aid and taxes.

Resources. With some of the most fertile land in tropical Africa and sufficient water resources, the Central African Republic has considerable agricultural potential. Diamond deposits produce more than one quarter of the country's total export earnings, and gold, uranium, iron ore, copper, and other minerals show promise.

Tropical rain forest covers large areas in the southwest of the country, and timber exports are a vital source of foreign exchange. The country's waterfalls are sources of hydroelectric power. The only major dams, located on the Mbali Lim River northwest of Bangui, produce most of the country's electricity.

Agriculture. Cultivation is largely the work of women, who grow a little cassava, some leafy vegetables, and a few bananas around their houses. In the savanna, peanuts, corn (maize), and some millet are still grown for family consumption. The main cash crops of coffee, cotton, and tobacco were introduced by French plantation owners, but, faced with political instability since independence, many Central African producers have returned to subsistence production. Newer cash crops being developed largely for import substitution purposes include oil palm and sugarcane.

The livestock population includes cattle, sheep, goats, pigs, and chickens, most of which are kept for domestic consumption. Pond-raised tilapia and river fish contribute substantial amounts of protein to the diet. The tsetse fly reduces the ranching possibilities for the Central African Republic, but development programs aimed at improving herds and migrations from Chad and The Sudan continue to increase the number of domestic animals.

Diamond mining

Industry. Diamonds are the main mineral product, but diamond mining activity has fluctuated as a result of disputes between the governing elite and a succession of foreign companies and because of smuggling. A small amount of gold is produced each year, but promised development of uranium, limestone, and other mineral resources has yet to be realized.

In comparison with neighbouring Cameroon, the Central African Republic's manufacturing sector is small; it is concentrated almost entirely in or near Bangui. Many sizable firms that once were the personal property of Emperor Bokassa were nationalized after his overthrow in 1979 and later sold to private owners.

The lack of space and poor service in Bangui hotels precludes mass tourism. Relatively exclusive tours to big game reserves in the far north are under foreign management; some passing trans-African expeditions are the only other major activities in the tourist sector.

Finance and trade. The government has experienced sizable budgetary deficits since the early 1980s. Supported by stand-by programs from the International Monetary Fund, direct budgetary aid from France, and assistance from other donors, the Central African government continues to struggle with the burden of a bloated and inefficient public sector. Foreign investment is theoretically welcomed and encouraged by liberal conditions for foreign investors and assistance to the private sector. Since import licenses are required for imports from outside the Franc Zone and payments for imports from countries outside the Franc Zone are subject to exchange control regulations, few non-French companies have sought to invest in the Central African Republic.

Three banks in the Central African Republic have more than 30 percent French ownership. The two commercial banks have expanding networks of provincial banks, and the Banque de Credit Agricole et de Développement supplies most of the nation's farm credit. The Central African Republic is a member of the African Financial Community (Communauté Financière Africaine; CFA); the currency is the CFA franc. The country is also an active member of the Central African Customs and Economic Union (UDEAC).

The Central African Republic relies heavily on its exports. The main exports—coffee, diamonds, timber, and cotton—have not produced enough revenue to achieve a positive trade balance since the mid-1970s. France is the country's leading trading partner, purchasing most of the coffee and tobacco produced. Belgium is also an important trading partner, purchasing most of the nation's diamond exports.

Transportation. With no direct access to the sea, no railways, and only about 300 miles (500 kilometres) of hard-surfaced roads, the country relies on its waterways for communication and commerce. More than 70 percent of the international trade is shipped by river. There are about 4,400 miles of inland waterways, though only some 750 miles are navigable year-round. The Ubangi-middle Congo route is the normal international transportation link with the outside world. This course is navigable most

Superstock

Boats moored along the banks of the Chari River, Central African Republic.

of the year from Bangui to Brazzaville, Congo, and from there goods are shipped by rail to the Atlantic port of Pointe-Noire.

Only the international airport, at Bangui-Mpoko, can take wide-bodied aircraft. There are several regional airports and many other airstrips, although internal services are irregular, depending on an unreliable supply of aviation fuel. The country has several international air links, including Air Afrique.

Administration and social conditions. *Government.* The territory known as the Central African Republic was granted political independence from France in August 1960; the current government is the result of a military coup in September 1981. This regime began to be somewhat liberalized after September 1985, when the military committee that had ruled under the direct guidance of General André Kolingba for four years was dissolved. At that time, Kolingba appointed a Cabinet with a number of civilian members. On Nov. 21, 1986, a new constitution creating a strong executive presidential form of government was adopted by referendum; on the same day, Kolingba was elected president. After the creation of a single-party state in February 1987, on July 31, 1987, for the first time in 20 years, members of a national assembly were elected for five-year terms by universal suffrage. In the spring of 1988, municipal elections for a slate of candidates presented by Kolingba's Central African Democratic Assembly were held. Since only candidates allowed by Kolingba may run for office, there is little opposition to executive decisions.

The country is divided into *préfectures, sous-préfectures* (subprefectures), and *communes.* Bangui is an autonomous commune.

Judiciary

Besides the Supreme Court, a court of appeals, and a criminal court, there are several lower tribunals and a military tribunal. Judges are appointed by the president. The judicial system is loosely based on that of France, with some vestigial traditional courts still operating on the local (subprefecture) level.

Education. In the late 1980s there were about 780 primary schools and 20 secondary schools in the country. The University of Bangui, founded by Jean-Bedel Bokassa in 1969, has operated fitfully since October 1970, offering courses in law, science, and medicine. In addition, the École Normale Supérieure, the National School of Administration, the National School of Arts, the Central School of Agriculture, and a number of other religious and technical schools offer students an education that is increasingly irrelevant for the needs of the desperately poor country. The best students, and especially those with the best political connections, continue to go to France for their education.

Health and welfare. For all practical purposes, no modern health care facilities exist outside of Bangui and a few other towns. A number of missionary-staffed and -run hospitals and clinics give relatively good care to those who can reach them. For the majority of Central Africans, however, the poorly equipped and poorly staffed maternity clinics, dispensaries, and first-aid posts theoretically available to them in the countryside are of little practical value. Even the Bangui hospital is far below minimal care standards; some private clinics are available to the wealthy in the capital. Sleeping sickness, malaria, leprosy, tuberculosis, and venereal diseases, including AIDS, are rampant. Data on the AIDS/HIV infection problem are not readily available, but the problem is growing and poses a major threat to the health of Central Africans.

Welfare benefits, including unemployment and maternity benefits, child-care allowances, and social security, are available to a small number of government and private-sector employees in the urban centres, but most people rely on their families and traditional village support systems for what little help they can get. The growing number of homeless youths in Bangui and in the other larger urban areas is a major social problem facing the nation.

Cultural life. The actual living situation of the majority of Central Africans has not been systematically studied. Neither the impact of the French colonial regime nor the unsettled state of affairs since independence has seriously altered the rural life patterns that have been evolving for centuries. In most Central African families women continue to play a crucial role in the gathering, production, conservation, distribution, and preparation of food. Hunting and fishing—male occupations by custom—remain important for the subsistence of many Central Africans, and the production of such commercial crops as coffee, cotton, and tobacco tends to be chiefly a male activity, but women are the principal food producers for household consumption.

Traditional crafts

Until the 19th century, craftspersons in the region produced many fine handmade items. The slave trade and the early years of colonization disrupted the expansion of crafts, however, and most of them disappeared. Today roughly woven mats and baskets, simple wooden utensils, and plain pottery are all that remain of traditional local handiwork. In recent years handicraft workers have begun producing unique designs and pictures made from butterfly wings glued to paper and some rather interesting ebony carvings.

Central African musicians form dance bands in Bangui and in some regional towns. These bands use electronic equipment to play popular music from around the world. Many groups also perform original Central African music, characterized by sad, haunting vocals and a strong, rhythmic beat.

Few works of literature from Central Africa have been published, although collectors are slowly gathering the traditional oral legends and folk stories from older villagers who remember them. Makombo Bamboté is the relatively well-known author of the novel *Princesse Mandupa* (1972) and *Coup d'état Nègre* (1987).

Central African artists have produced both watercolour and oil paintings. The murals and canvases of Jerome Ramedane depict scenes of African animal life, hunting parties, and daily village life. Similar works are often found on the walls of restaurants, bars, and other gathering places in Bangui and other towns. The Musée National Ethnologique Centrafricain (Musée Barthélemy Boganda) in Bangui contains exhibits of traditional musical instruments, implements of warfare, village architecture, hunting tools, pottery, and religious objects.

Football (soccer) and dancing are the major recreational pastimes of young Central Africans. Basketball and some rugby are also played, especially in Bangui, and those who can afford it may take in films in Bangui or one of the larger towns. Religious functions and traditional folksinging and dancing are also part of the yearly cycle for many Central Africans.

The national radio system broadcasts widely, but the national television system is limited mainly to the Bangui district. The local press is comparatively undeveloped, being based until 1986 on government news bulletins. The main publications are *Centrafric Presse,* a daily information bulletin; *Terre Africaine,* a weekly; and *Le Courrier Rural,* a regular publication from the Chamber of Agriculture. There is one government-controlled press agency, the Agence Centrafricaine de Presse. In 1986 the nation's first daily newspaper, *E le Songo,* began publication; it is also government-controlled. (J.v.H./T.E.O'T.)

For statistical data on the land and people of the Central African Republic, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Early history. Diamond prospectors in the Central African Republic have found polished flint and quartz tools at least 8,000 years old. About 2,500 years ago local farming people set up megaliths weighing several tons near Bouar. The cooperation necessary to make and position these monuments suggests that they were built by fairly large social units. By the 15th century AD, ethnic groups speaking languages related to those spoken by present-day Central Africans were living in the area. These ethnic groups lived in relatively isolated small settlements. They used iron tools and cleared land by the slash-and-burn method.

Emergence of ethnic groups

Until the 17th century, the region of the Central African Republic was not directly connected to external commercial routes. At that time, Arabic-speaking slave traders extended the trans-Saharan and Nile River trade routes into the region. Until the mid-19th century, Arabic-speaking and local African traders captured people and sent them as slaves to North Africa or down the Ubangi and Congo rivers to the Atlantic coast. From these points, slave ships transported the captives to the Americas.

By the mid-19th century, the Bobangi people from the Ubangi River area had become major slave traders themselves and raided for captives among the nearby Baya and Mandjia peoples. Many ethnic groups in Central Africa still harbour resentment against the ruling elite, who tend to come from riverine groups akin to the Bobangi. The disruption caused by the slave trade hampered the development of strong social organization among the region's ethnic groups.

The colonial era. Toward the end of the 19th century, faster ships, more powerful weapons, and the use of quinine against malaria made it possible for Europeans to gain control of vast areas of Africa. The Industrial Revolution in Europe had created the need for more markets and for new sources of raw materials. During the last two decades of the 19th century, Belgium, Great Britain, Germany, and France competed for control of what became the Central African Republic. By the beginning of the 20th century, frontiers had been established for the Ubangi-Shari (Oubangui-Chari) colony by the European powers. Many Africans resisted French control. Several military expeditions in the first decade of the 20th century were needed to crush their resistance.

To squeeze a profit from the colony, the French government leased large tracts of land to private European companies. Once leases were granted, firms had almost complete power to administer the land as they desired. Company overseers forced men to gather wild rubber and to work on plantations and public projects. At the same time, famine, disease, and the absorption of village lands into European plantations further disrupted the lives of Central Africans.

In the late 1920s, the French began to create a mobile health system in the colony to fight disease, especially sleeping sickness. Roman Catholic and Protestant missionaries set up a few schools and medical clinics. Through the use of forced labour, cotton cultivation was increased and a good network of roads was established. By the late 1930s, gradual increases in gold and diamond production began to produce a profit for French investors.

During World War II General Charles de Gaulle called on residents of French territories to help fight the Germans, and 3,000 Central Africans responded. After the war these troops returned to their homeland with a new sense of national, rather than simply ethnic, identity. After the war de Gaulle organized the French Union. He created new local assemblies, and French colonists and a few Africans became regional political representatives. On Nov. 10, 1946, Barthélemy Boganda became the first Central African elected to the French National Assembly.

Government of Barthélemy Boganda

Independence. A Roman Catholic priest, Boganda left the priesthood and formed the Social Evolution Movement of Black Africa (Mouvement pour l'Evolution Sociale de l'Afrique Noire; MESAN). In 1957 MESAN gained control of the Territorial Assembly, and Boganda became president of the Grand Council of French Equatorial Africa. Boganda hoped that the former French territories of Chad, Gabon, Congo, and Ubangi-Shari could form a single nation. When the others rejected the unification plan, Boganda reluctantly agreed to accept the new constitution offered to Ubangi-Shari by France.

After Boganda's death on March 29, 1959, David Dacko, a former teacher, became president. The Central African Republic was granted independence on Aug. 13, 1960. Dacko allowed France broad authority in trade, defense, and foreign relations. To reward his ethnic supporters, Dacko multiplied government positions, and the increased number of salaries drained the national budget.

In 1962 Dacko made MESAN the only legal national political party. In the elections of early 1964 Dacko ran unopposed and was, for the first time, formally elected as president. The economy declined and the national debt soared. On Dec. 31, 1965—amid impending bankruptcy and a threatened nationwide strike—the commander of the army, Jean-Bedel Bokassa, replaced Dacko in a rather uneventful coup d'état.

Bokassa abolished the constitution, dissolved the legislature, and turned over administrative duties to his appointed Cabinet. Bokassa allowed no opposition. Despite his increasingly dictatorial rule, France continued to support him and the country's faltering economy. The French wanted to retain control of the diamond—and potential uranium—output of the country. French paratroopers helped bolster Bokassa's rule in the mid-1970s.

Creation of the Central African Empire

In 1972 Bokassa declared himself president for life, and on Dec. 4, 1977, he renamed the country the Central African Empire and crowned himself Emperor Bokassa I with lavish ceremonies largely financed and organized with French government support. Under Bokassa's rule, the government went further into debt while the emperor personally administered the nation's diamond trade, keeping most of the profits. Finally, in September 1979 the French government removed Bokassa and restored Dacko as their client ruler.

Dacko's second turn at governmental control was not received with enthusiasm. To maintain his power, Dacko relied on French paratroops and on administrative officials who had served in Bokassa's government. As opposition grew, labour strikes and bomb attacks undermined Dacko's control. He increasingly relied on the army to retain power. On Sept. 1, 1981, General André Kolingba, a member of another riverine ethnic group, the Yakoma, removed Dacko from office in a relatively bloodless coup and established a military government.

The government stayed almost entirely in military hands until 1985. In that year Kolingba dissolved the military committee that had ruled the country since 1981 and named a new 25-member Cabinet with a few civilians. In early 1986, under pressure from the World Bank and other international organizations, Kolingba had a new constitution approved by the national assembly; it was adopted following a referendum later that year. Legislative elections were held in July 1987, but the government continued to operate under the direct control of Kolingba, who was elected in a 1986 referendum to six more years as head of state. As president, he effectively held all executive and legislative power in the nation. (T.E.O'T.)

For later developments in the history of the Central African Republic, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Congo

The People's Republic of the Congo (République Populaire du Congo)—commonly known as the Congo—lies astride the equator in west-central Africa. It is bordered to the west by Gabon, to the northwest by Cameroon, to the north by the Central African Republic, and to the east and south by Zaire. To the southwest it shares a border with the Angolan enclave of Cabinda; the republic also has a coastline 100 miles (160 kilometres) long on the Atlantic Ocean. The Congo has an area of 132,047 square miles (342,000 square kilometres), but the country as a whole is sparsely inhabited, with somewhat more than half of its population concentrated in the towns. The national capital of Brazzaville is an important inland port on the Congo River.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* The country is fringed by a narrow coastal plain 40 miles (64 kilometres) wide, which stretches for about 100 miles between Gabon and Cabinda. The plain rises gradually from the sea to the Mayombé Massif, a low mountain range that parallels the coast. The Mayombé peaks are quite sharp and are separated by deep river gorges. At the southern end of the range, Mount Foungouti attains 3,051 feet (930 metres). The northern peaks are lower; among them, Mount Moguindou rises to 2,132 feet (650 metres).

East of the Mayombé Massif lies the Niari valley, a 125-mile-wide depression. Toward the north the terrain rises gradually to the Chaillu Massif, which reaches elevations between 1,600 and 2,300 feet on the Gabon border; toward the south the depression rises to the Cataractes Plateau. The valley is an important passage route between the inland plateaus and the coast.

Beyond the Niari valley is a series of plateaus about 1,600 feet above sea level, separated by the deeply eroded valleys of tributaries of the Congo River. The Bembe Plateau lies between the Niari valley and the Chaillu Massif, while the Batéké Plateau stretches northward along the Congo River from Brazzaville to Mpouya.

The Congo basin

The northeast is composed of the western reaches of the Congo basin; from the western mountains and plateaus a vast 60,000-square-mile plain slopes eastward to the Congo River. Cut by numerous Congo tributaries, the plain is swampy and floods annually.

Drainage and soils. The country's drainage system is dominated by the Congo (Zaire) River. The Congo's main tributary, the Ubangi (Oubangui) River, flows southward from the Central African Republic and forms the country's eastern border until it reaches the town of Liranga, where it joins the Congo proper. The main river continues southward to Malebo (Stanley) Pool, a shallow 300-square-mile lake, and on to Livingstone Falls before turning southwest through Zaire to enter the Atlantic Ocean. The major right-bank tributaries of the Congo, all within the republic, include the Sangha, Likouala, Alima, Nkéni, Léfini, Djoué, and Foulakari rivers.

The coastal watershed is formed by the Kouilou River, which flows southwestward for about 450 miles from its source in the plateau region to Kayes, where it empties into the Atlantic. Through the Niari valley to Makabana, where it joins the Louessé River to form the Kouilou proper, it is called the Niari River. The stream is broken by numerous waterfalls, and the banks are irregular. The mouth is blocked to navigation by sandspits formed by the strong Benguela Current.

About two-thirds of the country is covered with coarse-grained soils that contain sand and gravel. Lateritic soils, with a high proportion of iron and aluminum sesquioxides, characterize low-lying areas. Because of the hot and humid climate, organic matter is decomposed by rapid bacterial action before it can accumulate into humus; moreover, topsoil is washed away by the heavy rains. In the savanna regions, the fertile alluvial soils are threatened with erosion by wind as well as rain. A diverse pattern of coarse- and fine-grained soils covers the plateaus and hills.

Climate. The tropical climate is characterized by heavy rainfall and high temperatures and humidity. The equator passes across the country just north of Liranga. In the north the dry season extends from November through March and the rainy season from April through October, whereas in the south the contrary is true. On both sides of the equator, however, local climates with two dry and two wet seasons may be found.

Annual rainfall is abundant throughout the country, but seasonal and regional variations are important. Precipitation averages more than 48 inches (1,200 millimetres) annually and often surpasses 80 inches.

Temperature range

Temperatures are relatively stable with little variation between seasons. Much greater variation occurs between day and night, when the difference between the high and low readings averages about 27° F (15° C). Over most of the country, annual average temperatures range between 68° and 81° F (20° and 27° C), although in the south the cooling effect of the Benguela Current may produce readings as low as 54° F (12° C). The average daily humidity is about 80 percent.

Plant and animal life. Nearly two-thirds of the Congo is covered with tropical rain forest. The dense growth of African oak, red cedar, walnut, softwood okoume, or gaboon mahogany, and hardwood limba (*Terminalia superba*) provides an evergreen canopy over the sparse undergrowth of leafy plants and vines. The coast and the swampy areas contain coconut palms, mangrove forests, and tall grasses and reeds. The plateau areas and the Niari valley are covered with grasses and widely spaced broad-leaved trees.

The forests contain several varieties of monkey, chimpanzee, gorilla, elephant, okapi, wild boar, and buffalo. Wildlife in the savanna regions includes several varieties of antelope, jackal, wild dog, hyena, and cheetah. On the plateaus, rhinoceroses and giraffes are numerous, but lions are scarce. Birdlife includes the predatory eagle, hawk, and owl, the scavenging vulture, and the wading heron.

Freshwater fish include perch, catfish, sunfish, and mudskippers. Crocodiles live throughout the Congo River. The numerous snakes include poisonous varieties: cobra, python, green mamba, and puff adder. Among the insects, the most dangerous are the tsetse fly, which causes sleeping sickness in human beings and a similar disease, called nagana, in cattle, and the mosquito, which carries malaria and yellow fever.

Settlement patterns. The country's four main regions developed around the historical concentrations of the main ethnic clusters. The southern region between Brazzaville and the coast is inhabited by the Kongo peoples. Also in the south, the Teke inhabit the Batéké Plateau region. In the north, the Ubangi groups inhabit the Congo River basin to the west of Mossaka, while the Binga Pygmies and the Sanga are scattered through the northern Congo basin. Precolonial trade between north and south stimulated both cooperation and competition, while French favouritism toward the peoples of the southwest enhanced interregional rivalries. Massive internal migration and urbanization following independence has, however, attenuated such conflicts.

Settlement of the Congo landscape is very uneven. The southwestern quarter of the country is home to 70 percent of the population. In the north and northeast, population is sparse. Congo is also highly urbanized by African standards; a majority of the population lives in cities, and because the urban growth rate far exceeds that of the country as a whole, urbanization continues to intensify. Since this growth has been chiefly the result of internal migration, most rural communities have ties to the larger national community and economy of the cities.

Urban centres

The major cities are Brazzaville, Pointe-Noire on the Atlantic coast, Nkayi (formerly Jacob) in the Niari valley, and Loubomo (formerly Dolisie) in the Mayombé region. Colonial creations by and large, the cities reflect French influence: a central administrative and commercial core is surrounded by residential areas. Before independence there was a marked difference between the spacious, planned European neighbourhoods and the less regimented, more populous African parts of town. Since 1960, however, greater social and economic mobility in the African population, attempts at urban renewal, and massive rural-to-urban migration have blurred these distinctions.

The people. *Ethnic and linguistic composition.* About half of the Congo's inhabitants belong to the Kongo peoples, whose major subgroups include the Sundi, Kongo, Lali, Kougni, Bembe, Kamba, Dondo, Vili, and Yombe. The Ubangi people include the Makoua, Kouyou, Mboshi, Likouala, Ngala, and Bonga. The Teke and the Sanga, or Gabonese Bantu, are also divided into various subgroups. The Binga Pygmies live in small bands, usually as clients of surrounding peoples. Most of the Europeans in the Congo are French who live in the main cities. There are also small populations of foreign Africans, Portuguese, and Chinese.

Except for the Pygmies, all the indigenous peoples speak their own Bantu languages. Intergroup communication and trade fostered the development of two trade languages, Lingala and Monokutuba. Lingala is spoken north of Brazzaville, and Monokutuba is common in the area between the capital and the coast. French is the official language and the medium of school instruction at all levels, as well as the language of the African upper class and the European community.

Independent churches

Religion. About half of the population practices classical African religions. The Christian community is composed of about two-thirds Roman Catholics and one-third Protestants, including members of the Evangelical Church of the Congo. There are also independent African churches: the Church of Jesus Christ on Earth by the Prophet Simon Kimbangu, the largest independent church in Africa, is a member of the World Council of Churches. Other independent churches include the Matsouana Church and the Bougist Church. Most of the small Muslim community is made up of foreigners who reside in Brazzaville or Pointe-Noire.

Demographic trends. Like many African countries, Congo has a fast-growing, relatively young population. Earlier in the 20th century, however, the country was part of the low fertility belt, a region stretching from Gabon to Uganda where many societies experienced little or no population growth. Life expectancy, among the lowest on the continent prior to 1950, has improved steadily in the last half of the 20th century and is now about average for mainland sub-Saharan Africa. Urban in-migration has long been an important demographic trend. During the colonial era, the new colonial cities, and Brazzaville in particular, attracted African migrants. Congo is today among the most urban countries of sub-Saharan Africa.

The economy. Petroleum and mining are the major export industries, followed by forestry and commercial agriculture. Light manufacturing (mostly shoes), sugar processing, and assembly industries also assumed greater importance in the 1980s. These activities, however, employed only a small fraction of the labour force, most of which was engaged in agriculture and the non-salaried informal urban economy.

In the late 1980s, following the fall in world oil prices, Congo experienced a major financial crisis. Negotiations for aid from the International Monetary Fund and the World Bank produced agreements to privatize portions of the national economy and to reduce the national bureaucracy. Such agreements may have improved the ability of Congo to compete in the international economy; at the same time, they did little to ameliorate the poverty of much of the population.

Resources. Important resources include petroleum and natural gas. Large reserves of potash (potassium chloride) are found at Tchitondi (Holle), 30 miles northeast of Pointe-Noire. Iron ore is found in the south and in the western Sangha basin. Minor deposits of gold and diamonds are located in the Kouilou valley, and there are copper and lead deposits west of Brazzaville. Oil reserves are found in the coastal zone. There are also deposits of zinc, tin, uranium, bauxite, and titanium (a metallic element used in alloys).

Forests of both softwoods and hardwoods cover nearly two-thirds of the country. The rivers and lakes are home to important fish resources. Power resources consist of petroleum, wood, charcoal, and the hydroelectric potential of Congo's rivers, which generate nearly all of the country's electricity.

Agriculture, forestry, and fishing. For the most part, agriculture is subsistence in nature and occupies about three-fifths of the work force. Poor soil and the lack of fertilizers produce low yields; the country is not self-sufficient in food production. Most of the cultivated land is in family holdings that are too small for mechanized farming; international development strategies, which are shaped by reliance on large-scale production, have yet to devise plans to reach such small-scale cultivators. In the savanna, land is cleared by burning, and women work the fields with hand tools. Cassava (manioc) is the basic food crop everywhere but in the south, where bananas and plantains are prevalent. Rice is grown in the Niari valley and in the north around Djambala. The diet is supplemented with yams, taros, sweet potatoes, corn (maize), peanuts (groundnuts), and fruit. Livestock consists of sheep, goats, pigs, and poultry. The government has sponsored the raising of cattle since the introduction in the 1960s of *n'dama* cattle—a breed resistant to the tsetse fly.

Commercial agriculture

The major cash crops are cacao, coffee, peanuts, palm kernels, tobacco, and sugarcane. All are grown in modest amounts. Other cash crops include rice, bananas, and cotton. Commercial agriculture, including cattle ranching, is concentrated in the Niari valley.

Forestry products accounted for more than 60 percent of the total exports in the late 1960s. Twenty years later, petroleum made up more than 90 percent of exports. Although forest reserves are extensive, inadequate transportation and flood conditions limit exploitation in the north; hence forestry operations are concentrated in the south. The Congo is the world's largest producer of limba and (after Gabon) the second largest producer of okoume. Products include logs, sawn wood, and veneers. Forestry was largely under French control until the 1960s, when African participation began to increase.

Commercial marine fishing is conducted off Pointe-Noire. The catch includes tuna, bass, soles, and sardines. Freshwater fishing on the rivers, lakes, and swamps is largely a subsistence activity.

Industry. Potash is mined, and petroleum is the most important export. Copper, zinc, lead, and iron ore are also extracted on a small scale.

The small manufacturing sector is hampered by limited domestic markets, dependence upon foreign investment, and the lack of skilled labour. Most factories are located in Brazzaville, Pointe-Noire, Kayes, Loubomo, and cities in the Niari valley. Products include processed foods, beverages, cigarettes, textiles and clothing, footwear, processed wood and paper, chemicals, cement and bricks, glassware, and metal goods such as nails and metal furniture. The first petroleum refinery went into operation in 1976 at Pointe-Noire. Handicrafts produced by individual artisans include carvings, pottery, needlework, tiles, and bricks.

Transportation. The road system is most developed in the south. Major routes link Brazzaville with Pointe-Noire and Loubomo with the Gabon border. Many roads are impassable during the rainy season.

Railways are also concentrated in the south. The major Congo-Ocean Railway line runs for about 320 miles from Brazzaville west through Nkayi and Loubomo to Pointe-Noire. There is also a 175-mile branch line from Favre north to Mbinda on the Gabon border. These railways offer important transshipment services for neighbouring countries, contributing significant revenues. They are also important to mining and industrial development, for most industrial towns are located along them.

Water transportation has long linked Congo, Chad, and the Central African Republic. The rivers, however, are often broken by rapids and are subject to seasonal variations in flow. Brazzaville is the most important inland port. Linked to Kinshasa, Zaire, by ferry, it is there also that passengers and freight traveling downriver from Bangui, in the Central African Republic, are transshipped overland by rail to the ocean port of Pointe-Noire. This seaport is the major transshipment centre for these three countries as well as western Cameroon.

Administration and social conditions. *Government.* Congo is a single-party people's republic led by the Congolese Labour Party (Parti Congolais du Travail; PCT). The party's 75-member Central Committee may initiate revision of the constitution (adopted in 1979 and amended in 1984), and it appoints members of the Revolutionary Court of Justice, which heads the judicial system. The Central Committee's chairman is the president of the republic, who is head of state and of government and who is elected by the party congress to serve a five-year term. The 133 members of the People's National Assembly (Assemblée Nationale Populaire) are elected by direct universal suffrage from a list of candidates chosen by the PCT; they also serve five-year terms. The judicial system consists of a Supreme Court, a court of appeals, and criminal, labour, and common law courts.

Role of the PCT

Congo is divided into nine *régions* and the capital district of Brazzaville, each headed by appointed district leaders and an elected council; it is further subdivided into *communes* and villages. In addition to Brazzaville, there are five other *communes*—Pointe-Noire, Loubomo, Mossendjo, Nkayi, and Ouesso—and eight administrative control posts subordinate to their regions. The village remains the primary administrative and political unit outside the cities.

Education. Education is free and compulsory for students between the ages of 6 and 16. The six-year primary education course includes instruction in agriculture, manual skills, and domestic science. On the secondary level courses are offered in vocational training, academic and technical training, general education, and teacher training. Institutions of higher learning include the Marien

Ngouabi University in Brazzaville and colleges and centres for specialized and technical learning. Congo enjoys a literacy rate much higher than most countries in sub-Saharan Africa.

Health and welfare. The most common health problems are respiratory diseases, malaria, tuberculosis, and intestinal parasites—all preventable maladies. Other diseases include trypanosomiasis (sleeping sickness), yellow fever, leprosy, yaws, and AIDS. Disease control is difficult because most water sources are polluted and sanitation is poor, even in the cities. The country's two largest hospitals are in Brazzaville and Pointe-Noire. Other health facilities include regional health centres, infirmaries, dispensaries, maternal and child-care centres, and private clinics. Mobile health units combat communicable diseases in remote areas.

Welfare services provided by the government, the labour union, and employers are largely limited to wage earners in the formal sector and their families. Benefits include old-age pensions, life insurance, worker's compensation, and family-allowance payments. Government-sponsored social workers operate among the poor.

Cultural life. Precolonial artistic expression emphasized ceremonial music, dance, sculpture, and oral literature. Christianity and colonialism had a great impact on these art forms. The carving of ritual objects became commercialized, and music and dance altered as a result of the introduction of Western instruments and musical styles. In the 1980s the Brazzaville region, along with Kinshasa, across the river in Zaire, became a vital centre of contemporary African music. There are two libraries in Brazzaville, and a national museum contains collections of prehistoric objects as well as precolonial and contemporary art. (A.Fu./D.D.C.)

For statistical data on the land and the people of Congo, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Early history. Human habitation of Congo came relatively late in the Sangoan era (100,000 to 40,000 BP), perhaps because of the dense forest. The people who used the large-core, bifacial Sangoan tools probably subsisted by food gathering and digging up roots; they were not hunters. Refined versions of this tradition continued through the Lupemban (40,000 to 25,000 BP) and Tshitolian eras. Only late in the first millennium did agriculture emerge in the savanna adjacent to the lower Congo River.

The early inhabitants were farmer-trappers, fishing peoples, and Pygmy hunters. People lived in households including kin and unrelated individuals; at the centre of the household was a "big man," who represented the group. Mobility—of individuals, groups, goods, and ideas—figured prominently and created a common social environment. Such intercommunication is evident from the closely related Bantu languages of the region. Speakers of the Eastern (Ubangian) languages lived in the north but maintained ties with their forest neighbours.

Emergence of three kingdoms

Larger-scale societies emerged between AD 1000 and 1500; they were based on clans whose members lived in different villages, village clusters with chiefs, and small forest principalities. Chiefdoms on the southern fringes became more complex; three kingdoms eventually developed: Loango, at the mouth of the Kouilou River on the Atlantic coast; Kongo, which had its beginnings in the first millennium, in the far southwest; and Tio, which grew out of small chiefdoms on the plains north of Malebo Pool. Rulers derived power from control over spirit cults, but trade eventually became a second pillar of power.

In 1483 the Portuguese landed in Kongo. Initially, relations between the Kongolese and Portuguese rulers were good. Characterized by the exchange of representatives and the sojourn of Kongolese students in Portugal, this period was a harbinger of late 20th-century technical assistance. Unfortunately, the need of Portuguese planters on São Tomé for slaves undermined this amicable arrangement by the 1530s.

Between 1600 and 1800, the slave trade expanded. Local leaders challenged state control. Among the Tio the western chiefs became more autonomous. Contact with Europeans also introduced new American food crops. Corn and cassava allowed greater population densities. This, along with the emergence of a "market" for foodstuffs, led to greater use of slaves, intensified women's work, and changed the sexual division of labour.

Role of Brazza

The colonial era. By the early 19th century the Congo River had become a major avenue of commerce between the coast and the interior. Henry Morton Stanley, a British journalist, explored the river in 1877, but France acquired jurisdiction in 1880 when Pierre Savorgnan de Brazza signed a treaty with the Tio ruler. The formal proclamation of the colony of French Congo came in 1891. Early French efforts to exploit their possession led to ruthless treatment of the local people, and Brazza returned in 1905 to lead an inquiry. In 1910 the French joined Congo with neighbouring colonies as the federation of French Equatorial Africa, with its capital at Brazzaville.

The French were preoccupied with acquiring labour. Forced labour, head taxes, compulsory production of cash crops, and draconian labour contracts forced Africans to build infrastructure and to participate in the colonial economy. No project was more costly in African lives than the Congo-Ocean Railway, built between 1921 and 1934 from Pointe-Noire to Brazzaville; between 15,000 and 20,000 Africans died.

In 1940 Congo rallied to the Free French forces. Charles de Gaulle, Governor-General Félix Eboué, and African leaders held a conference in Brazzaville in 1944 to announce more liberal policies. In 1946 Congo became an overseas territory of France with representatives in the French Parliament and an elected Territorial Assembly. Ten years later, the *loi cadre* (enabling act) endowed the colony with an elected government. Congo became a republic within the French Community in 1958 and acquired complete political independence in August 1960.

Congo since independence. Two major parties existed at independence: the African Socialist Movement (Mouvement Socialiste Africain; MSA) and the Democratic Union for the Defense of African Interests (Union Démocratique pour la Défense des Intérêts Africains; UDDIA). The two parties pitted the north against the south, an opposition that stemmed from the privileged place occupied by the southern Kongo and Vili in the colonial era. The two parties also had different political philosophies. The MSA favoured a powerful state and a partially publicly owned economy; the UDDIA advocated private ownership and close ties with France. UDDIA leader Fulbert Youlou formed the first parliamentary government in 1958; in 1959 he became premier and president.

Corruption, incompetence, mass disapproval, general strikes, and lack of French support led to Youlou's ouster in 1963. His successor, Alphonse Massamba-Débat, shifted policies to the left, notably by founding the National Revolutionary Movement (Mouvement National de la Révolution; MNR) as the sole party. The country sought assistance from the Soviet Union and the People's Republic of China and voted with the more radical African states in world forums. Regionally, Congo extended concrete support and offered a geographic base for the Popular Movement for the Liberation of Angola (MPLA), the Marxist movement that won independence for that country. Congo also offered asylum to the Patrice Lumumba followers who fled neighbouring Zaire and plotted for a return to power there.

Creation of the People's Republic

Regionalism and policy failures led the military to replace Massamba-Débat with Major Marien Ngouabi in 1968. Ngouabi maintained a socialist line, renaming the country the People's Republic of the Congo on Dec. 31, 1969; the Congolese Labour Party (PCT) replaced the MNR at the same time. Ngouabi was a northerner, and his regime shifted control of the country away from the south. Such moves created opposition among workers and students in the highly politicized environment of Brazzaville and other southern urban centres. Ngouabi was assassinated in March 1977. His successor, the more conservative Colonel Joachim Yhombi-Opango, soon clashed with the PCT, and Colonel Denis Sassou-Nguesso replaced Yhombi-Opango in 1979.

Although Sassou-Nguesso represented the more militant wing of the PCT—and immediately introduced a new constitution intended as a first step toward building a Marxist-Leninist society—he improved relations with the Western nations and with France. The regime's political language became more moderate, but the inefficient state enterprises created by earlier socialist policies remained in place in the early 1980s. They had been subsidized by petroleum production, but the drop in oil and other raw material prices led to economic crisis. The external debt surpassed $1.5 billion in 1985, and debt service consumed 45 percent of state revenue. Negotiations with the International Monetary Fund the following year led to an agreement to help the national economy in exchange for cuts in public spending and the state bureaucracy. At the end of the 1980s, however, the debt had tripled in size and the economic crisis continued. (D.D.C.)

For later developments in the history of Congo, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Gabon

The Gabonese Republic (République Gabonaise), on the west coast of Africa, lies astride the equator, with a total area estimated at 103,347 square miles (267,667 square kilometres). The republic is bordered by Equatorial Guinea and Cameroon to the north, the People's Republic of the Congo to the east and south, and the Atlantic Ocean to the west; the island state of São Tomé and Príncipe is situated off the coast. Gabon's capital is Libreville. Gabon remains strongly attached to France, its former colonizer, and to the French language and culture.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief and drainage.* The narrow coastal plain—often no more than 20 miles (32 kilometres) wide in the south—is formed of sandstone and alluvium; northward it broadens to a width of 100 miles, with outcrops of chalk, limestone, and Cretaceous sandstone. North of the most westerly point, Cape Lopez, the contour of the coast becomes more jagged. The Atlantic's northward-flowing Benguela Current, which softens the southern coastline by creating sandbars, is no longer effective here. Inland, the relief is characterized by a series of granite plateaus, extending generally northwest to southeast and rising to heights of from 1,000 to 2,000 feet (300 to 600 metres). Farther to the west and north, the Cristal Mountains have been dissected by the river system from the western plateau escarpment into a distinct upland area. To the south, the Ogooué River drains through a sandstone saddleback before descending to the lowlands through the granite formations of the Lambaréné region. Granite also forms the Chaillu Massif, Gabon's central watershed, south of the Ogooué, which rises to an elevation of approximately 3,000 feet, except where topped by the 3,215-foot (980-metre) elevation of Mount Iboundji.

Climate. Gabon has an equatorial climate, with year-round high temperatures and humidity. Rainfall varies from an annual average of 120 inches (3,050 millimetres) at Libreville to 150 inches on the northwest coast, with almost all of it falling between October and May. In the period from June to September there is little, if any, rainfall, but humidity remains high. Temperature shows little seasonal variation, the daily average being about 81° F (27° C).

Plant and animal life. About 75 percent of the country is covered by a dense equatorial rain forest containing more than 3,000 species of vegetation, including the okoume, a hardwood tree that forms the backbone of Gabon's wood industry. The rain forest is inhabited by antelope, monkeys, gorillas, numerous tropical birds, and several varieties of elephants. (J.v.H./D.E.G.)

The towns

Settlement patterns. Gabon's major towns—Libreville, Port-Gentil, Lambaréné, Moanda, Mouila, Oyem, Franceville, and Bitam—contain nearly two-fifths of the total population. The remainder is scattered widely among some 4,000 rural villages, which are concentrated along the rivers and roads, often with a population of no more than a few families. Port-Gentil is the centre of the wood and petroleum industries, and Libreville is the administrative capital and commercial centre.

The people. *Ethnic and linguistic composition.* All of Gabon's 40 or so ethnic groups, except the few thousand Pygmies, speak Bantu languages and, on that basis, can be classified into 10 larger groups. The Myene group (including the Mpongwe and Orungu), though only a small part of the population today, has played an important role in the history of the country as a result of its location along the northern coasts. The Fang, who belong to the larger Ewondo or Pahouin group also found in southern Cameroon and Equatorial Guinea, account for one-third of the population; they live north of the Ogooué River. The largest groups south of the Ogooué are the Sira (including the Eshira and Punu), the Nzebi (Njabi), and the Mbete; these groups together form close to half the population. Less numerous peoples include the Benga and Seke (Sheke) in the far northwest, the Kota and Teke in the east, and the Vili along the far southern coast.

Many of the Bantu languages do not have written forms. During the 19th century Christian missionaries transliterated several of them in the Latin alphabet and prepared Bible translations and catechisms for their followers. But the French policy of limiting the use of indigenous languages solely to religious instruction inhibited the growth of other types of literature. Because of the extensive efforts to teach French, at least one-third of the Gabonese can speak the language, and more than one-quarter can read it.

Religion. A large majority of Gabon's population is Christian, with about three times as many Roman Catholics as Protestants. Though Gabonese serve as Roman Catholic bishops, they rely heavily upon foreign clergy, particularly the French Holy Ghost Fathers. The largest Protestant body, the Evangelical Church of Gabon, has Gabonese pastors in its parishes throughout the north. There also exist a small but growing Christian Alliance Church in the southwest and the tiny Evangelical Pentecostal Church (Assembly of God) in the estuary and far northern regions. A syncretic religion called Bwiti (based on an earlier secret society of the same name) came into being in the early 20th century and later played a role in promoting Fang solidarity. Most of the few thousand Muslims are immigrants from other African countries.

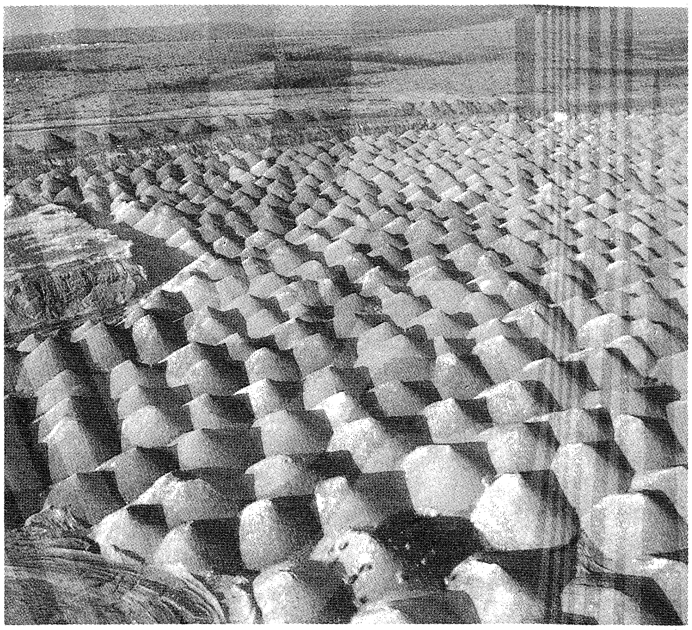
Demographic trends. Gabon, like its central African neighbours, has a very low population density. Since 1970, as a result of increased urbanization, the low rate of natural increase of the previous half-century appears to have given way to a relatively high growth rate. The extent to which the heavy immigration of foreign workers and Equatorial Guinean refugees has contributed to this growth is unclear. The population is relatively young—half are below age 22. Life expectancy, at 50 years for men and 53 years for women, is about average for the continent.

Labour shortage

The economy. Gabon's economy has more links with European and American markets than with those in neighbouring states (with the exception of Cameroon) or elsewhere in Africa. The economy shares some of the characteristics of those of other tropical African states: strong links with the former colonial ruler, a large degree of foreign investment and control, dependence on foreign technicians, and the decline of agriculture. Gabon differs from these states in its reliance on thousands of wage earners from other African countries to supplement its own sparse supply of workers in retailing, artisanship, and domestic transport.

Resources. Gabon possesses important resources in woods and minerals and much hydroelectric potential. But its poor transportation infrastructure and lack of financing, as well as unfavourable world market conditions, hinder the development of some of these resources.

Agriculture, forestry, and fishing. Although agriculture (mainly subsistence farming) occupies about two-thirds of the population, it plays but a small part in the economy of the country as a whole. Moreover, its appeal as a way of life has declined. Better educational and employment opportunities in the towns and cities have emptied the countryside of young people. Despite government efforts during the 1970s to promote development that would



A manganese mine in Moanda, Gabon.
Bernard Regent/The Hutchison Library

stem the rural exodus and raise foodstuffs for urban markets, by 1980 Gabon was producing only enough food to satisfy 10 to 15 percent of its needs. During the 1980s the government turned to expensive capital-intensive projects for market gardening to supply Libreville and Franceville. Efforts to revive cocoa and coffee production brought only modest results, but new projects for sugar refining at Franceville and palm oil processing at Lambaréné have worked out well. The prevalence of the tsetse fly has defeated attempts to raise beef and dairy cattle, though chicken raising has succeeded. Commercial fishing, though having considerable potential, is little developed.

For many years Gabon's forests, covering 75 to 85 percent of its territory, were its principal natural resource. Newly discovered and exploited mineral wealth had by the early 1970s forced raw wood and forest products into second place, and by the early 1980s, wood accounted for less than 10 percent of Gabon's exports. The principal forest districts have been at Kango, Booué, Fougamou, Ndjolé, Mitzic, and Mouila. The forest resources near the coast and rivers have been largely depleted. Exploitation of the forests of the interior began in the late 1970s with the construction of the first section of the Transgabon Railroad.

Industry. Gabon is the world's fourth largest producer of manganese; it possesses an estimated 25 percent of the world's known reserves outside the Soviet Union. Expansion of production at Moanda has been possible since the completion of the railroad to nearby Franceville in December 1986 and by the completion of improved ore handling facilities at the rail terminus at the deepwater port of Owendo in 1988. The exploitation and processing of uranium, which began in 1961 16 miles north of Moanda, mainly serves the needs of the French Atomic Energy Commission. Reserves estimated at 35,000 tons will last until about 2030 at present production levels. Reserves of nearly one billion tons of high-quality iron ore (60–65 percent iron content) discovered in the northeast at Mékambo in the 1950s have not been exploited because of the lack of transportation facilities. Plans to run a section of the Transgabon Railroad from Booué to this area were postponed because of lack of financing and depressed world market prices.

Mining industries

Since the late 1960s, revenues from petroleum have brought the government of Gabon unprecedented income, which it has used to construct infrastructure and to fund the expansion of education and health services. National budgets multiplied 15 times between the late 1960s and late '70s, when petroleum came to represent 70 percent of the country's exports. Despite slumping prices and the resulting drops in production, revenues from petroleum still provided 65 percent of national budgets during the late 1980s. More than 80 percent of production is from offshore fields, which are the most productive near Port-Gentil. The major onshore production sites are at Sette Cama and Rabi-Kounga. Gabon exports 85 percent of its petroleum production outside Africa, with the bulk of the crude oil going to France, the United States, Brazil, and Argentina. Natural gas from the fields at Port-Gentil is used largely to generate electricity.

Light industry has been expanding and diversifying since the opening in 1967 of a petroleum refinery at Port-Gentil. The refinery and its support operations (a shipyard and metalworking facilities) overshadow other manufacturing enterprises, which include lumber processing, cement and cigarette factories, a sugar refinery, breweries, light electronics, palm oil and flour mills, and textile-printing factories.

Finance and trade. Membership in the Franc Zone gives Gabon considerable stability. The CFA (Communauté Financière Africaine) franc is automatically convertible to French francs, giving trading partners confidence in Gabonese currency. The government has also encouraged foreign investors with its policy of economic liberalism, although there is governmental direction and planning.

France is Gabon's most important trading partner, receiving more than a third of its exports and providing half of its imports. Other important partners are other European Common Market nations, the United States, and Japan. These same countries provide the bulk of investment funds and foreign assistance; France alone is the source of more than one-half of the total. Saudi Arabia has also given financial help. Gabon belongs to a customs union with five other central African states and a common market with four others.

Transportation. The lack of good transportation facilities has long hindered Gabon's development. The Ogooué River is navigable from the Atlantic to Ndjolé, 150 miles upstream. The Ogooué and such rivers as the Abanga and the Nyanga can be used to float logs downstream from the interior.

Infrastructure development

The difficulty of building and maintaining all-weather roads led to an expansion of air transport after World War II. Gabon acquired a network of airfields served by light planes as well as international airports at Libreville, Port-Gentil, and Franceville. But air transport could not move such bulk goods as timber and minerals. Thus in the 1970s petroleum revenues were used to construct the Transgabon Railroad, or Transgabonais, to move such products and to prepare for the time when petroleum reserves would be depleted. With loans and aid from France, West Germany, and international organizations, work began in 1974. The first section, from Owendo to Ndjolé, opened in 1979; the second section, to Booué, in 1983; and the third, to Franceville, at the end of 1986.

Administration and social conditions. *Government.* Under the frequently revised constitution of 1961, Gabon is a republic under the executive direction of a president elected by direct universal suffrage for a period of seven years and a Council of Ministers appointed by the president. Provision is also made for a prime minister (appointed by the president) and a National Assembly (elected by direct suffrage) to assume legislative responsibilities and for an independent Supreme Court. After the French intervention of 1964, power became concentrated in the presidency. In 1968 the president's party, the Gabon Democratic Party (Parti Démocratique Gabonais; PDG) was declared the only one legally permitted. But in May 1990 transitional constitutional arrangements reestablished a multiparty system. A committee was appointed to formulate a new constitution.

The judicial system consists of a series of customary law courts at the lowest level, above which are a criminal court, a court of appeals, and the Supreme Court (all located in Libreville). The High Court, which is composed of members elected from the National Assembly, has the power to try the president and members of government. An audit office (*cour des comptes*) was created in 1977 to oversee the government's finances.

Administratively, Gabon is divided into nine *provinces,* which are further divided into *préfectures* and *sous-préfectures* (subprefectures). The provincial governors, the prefects, and the subprefects are all appointed by the president.

Education. The educational system continues to be modeled closely after that of France. French remains the sole medium of instruction; Bantu languages are studied as electives at the secondary and higher levels. While education is officially mandatory from the ages of 6 to 16, the bulk of children do not attend long enough to achieve literacy or numeracy. The Omar Bongo University, founded in 1970, has two- and three-year programs in most fields and some advanced studies. The University of Science and Technology of Masuku, near Franceville, opened in 1986. Many Gabonese study abroad, particularly in France, at the university and graduate levels.

Health and welfare. Health facilities remain inadequate, particularly outside the Libreville area, despite improvements during the 1970s and '80s. The government provides nearly all health care services. The internationally known hospital operated by Albert Schweitzer at Lambaréné from 1924 to 1965 receives government assistance. Malaria, sleeping sickness, tuberculosis, syphilis, and other infectious diseases are widespread problems.

Influence of French culture

Cultural life. A great deal of the cultural life of Gabon continues to be derived from or influenced by France. Gabon's contemporary writers express themselves almost exclusively in French. At the same time, there has been continued interest in the precolonial history and traditions of Gabon's peoples. Examples are the research on the Fang epic (*mvet*) and the art of the Mpongwe, Fang, and Kota. In 1983 the International Centre for Bantu Civilizations (Centre International des Civilisations Bantu; CICIBA) was created, with its headquarters at Libreville.

Gabon has newspapers in French, national and provincial radio stations broadcasting in French and local languages, and television broadcasting from Libreville that can be received as far away as Lambaréné. French publications circulate extensively, and television programs are relayed from France. (B.Wn./D.E.G.)

For statistical data on the land and people of Gabon, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Early colonization. Little is known about the history of the Gabon region prior to the arrival of the first Portuguese navigators in the Gabon Estuary in 1472, for the Bantu-speaking peoples inhabiting the region lacked writing. At that time portions of southern Gabon were loosely linked to the state of Loango, which in turn formed a province of the vast Kongo kingdom. From the offshore islands of São Tomé and Príncipe, where the Portuguese established sugar plantations, they developed trade with the mainland. From the late 1500s, Dutch, French, Spanish, and English competitors also exchanged cloth, iron goods, firearms, and alcoholic beverages for hardwoods, ivory, and a few slaves.

The slave trade

The slave trade achieved extensive development only between the 1760s and 1840s, as a result of heightened demand from Brazil and Cuba. Interior peoples sent undesirables from their own societies and captives from warfare down the waterways to the coast where they were collected in barracoons (temporary enclosures) to await the arrival of European ships. The Orungu clans at Cape Lopez organized a kingdom whose power rested on control of the slave trade through the mouths of the Ogooué River. The Mpongwe clans of the estuary, who were already important traders, also profited from the slave trade, as did the Vili of Loango, whose activities extended throughout southern Gabon. Only the Fang, who were migrating southward from Cameroon into the forests north of the Ogooué, ordinarily refused to hold slaves or engage in warfare to obtain them. The coastward migrations of the numerous and often warlike Fang nevertheless contributed to the further decimation and dispersion of many interior peoples, particularly during the 19th century.

French control. By 1800 the British were becoming the leading traders in manufactures throughout the Gulf of Guinea. After 1815 the French sought to compete more actively in the commercial sphere and to join Britain in combating the slave trade. To these ends, Captain Édouard Bouët-Willaumez negotiated treaties with the heads of two Mpongwe clans, King Denis (Antchouwe Kowe Rapontchombo) on the southern bank of the estuary in 1839 and King Louis (Anguile Dowe) on the northern bank in 1841. They agreed to end the slave trade and to accept French sovereignty over their lands. The arrival of American Protestant missionaries on the northern bank in May 1842 to open a school in the lands of King Glass (R'Ogouarowe)—the centre of British, American, and German commercial activity—spurred the French to establish Fort d'Aumale within the territory of King Louis in 1843. In 1844 France brought in Roman Catholic missionaries to promote French cultural influence among the Mpongwe and neighbouring peoples. French agents obtained a treaty from King Glass recognizing French sovereignty. In 1849 Bouët-Willaumez organized a small settlement of mainly Vili freed slaves called Libreville ("free town"), which with the fort formed the nucleus of the capital.

During the 1850s and '60s the French gradually extended their control along the adjacent coast and sent explorers into the interior. The expeditions of Pierre Savorgnan de Brazza between 1875 and 1885 established French authority on the upper Ogooué, where Franceville was founded in 1880, and on the Loango coast. An enlarged Gabon was attached to the French Congo in 1886 under Brazza as governor.

Establishment of frontiers

In 1910 Gabon became one of the four colonies within the federation of French Equatorial Africa. The French delimited the frontier with the Germans in Cameroon in 1885 and with the Spanish in Río Muni, or Spanish Guinea (later Equatorial Guinea), in 1900. French occupation of the Gabon interior brought little opposition; but interference with trade and such exactions as a head tax, a labour tax for public projects, and forced labour provoked considerable resistance, as did the French policy from 1898 to 1914 of seeking to develop the economy through monopolistic concessionary companies, which devastated settlement, agricultural production, and trade.

The period between the two world wars saw the creation of a pro-French but anticolonialist elite, mainly from the graduates of the boys' schools of the Brothers of Saint-Gabriel at Libreville and Lambaréné. From their ranks came most of the politicians who held office during the Fourth French Republic (1946–58) when Gabon became an overseas territory with its own assembly and representation in the French Parliament. In this era France considerably expanded public investment in the economy, in health care, and in education. In 1958 Gabon became an autonomous republic within the French Community and, after concluding cooperation agreements with France, achieved independence on Aug. 17, 1960.

Gabon since independence. Gabon favoured close relations with France and continued use of the French language and culture. It opposed political ties with the other states of Equatorial Africa, however, because of dissatisfaction with the previous federation and a desire to develop its natural resources for its own benefit.

M'ba's presidency

Attempts by the republic's first president, Léon M'ba, to institute a single-party regime provoked a rebellion by young military officers in February 1964. But M'ba, who had strong backing from French economic interests, was restored to power by forces sent on orders from General de Gaulle. This action further weakened the opposition forces led by Jean-Hilaire Aubame, previously deputy in Paris during the Fourth Republic. The intervention made possible the rise of Albert-Bernard (later Omar) Bongo to the presidency after M'ba's death in 1967 and the establishment of a single-party regime. In 1982 a new opposition group, the Movement for National Renewal (Mouvement de Redressement Nationale; MORENA), called for multiparty democracy, exercise of civil liberties, and an end to governmental corruption, but it was quickly suppressed.

In 1985 declining petroleum prices caused an economic downturn. Austerities imposed as part of a structural adjustment set off popular demonstrations against the Bongo government at Libreville and Port-Gentil in 1990. This agitation spurred the creation of a national conference that included opposition groups. The death of an opposition leader in May 1990 sparked violent disorders that led to French military intervention to protect French nationals and property. Opposition groups won seats in the new legislative assembly elected in September, but irregularities nullified the results in half of the 120 districts. (D.E.G.)

For later developments in the history of Gabon, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Rwanda

Lying east of Lake Kivu in the heart of the Western Rift Valley, the Republic of Rwanda (French: Rêpublique Rwandaise; Rwanda: Republika y'u Rwanda) is bounded on the north by Uganda, on the east by Tanzania, on the south by Burundi, and on the west by Zaire. Like its neighbour to the south, Rwanda is a country of minute dimensions (10,169 square miles [26,338 square kilometres]), grinding poverty, and high population density. It also shares with Burundi a long history of monarchical rule. Unlike what happened in Burundi, however, the demise of the kingship came about through a grass-roots, Hutu-led upheaval that reached its denouement before the country became independent. In Rwanda the state was forged from the ground up, in the crucible of a peasant revolution; in Burundi, by contrast, it is the state that seeks to forge society from the top down. Thus, if both states can be described as ethnocracies, in Rwanda the term carries a distinctly democratic connotation inasmuch as the Hutu represent the overwhelming majority of the population.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* The landscape is reminiscent of a tropical Switzerland. Its dominant feature is a chain of mountains of rugged beauty that runs on a north-south axis and forms part of the Congo-Nile divide. From the volcanoes of the Virunga Mountains in the northwest—where the Karisimbi reaches 14,787 feet (4,507 metres)—the altitude drops to 4,000 feet (1,220 metres) in the swampy Kagera (Akagera) River valley in the east. The interior highlands consist of rolling hills and valleys, yielding to a low-lying depression west of the Congo-Nile divide along the shores of Lake Kivu.

Principal rivers

Drainage. Except for the Ruzizi, through which the waters of Lake Kivu empty into Lake Tanganyika, most of the country's rivers are found on the eastern side of the Congo-Nile divide, with the Kagera, the major eastern river, forming much of the boundary between Rwanda, Burundi, and Tanzania.

Soils. The best soils, formed from volcanic lavas and alluvium, are found, respectively, in the northwest and along the lower portions of the larger river valleys. Elsewhere the largely metamorphic bedrock has produced soils of generally poor quality. The combination of heavy rainfall and deforestation has set in motion a process of extreme soil erosion that requires a burdensome investment of time and energy to curtail.

Climate. Elevation accounts for Rwanda's generally mild temperatures, which average 70° F (21° C) year-round at Kigali, for example, in the interior highlands. There are significant variations, however, between the region of the volcanoes in the northwest, where heavy rainfalls are accompanied by lower average temperatures, and the warmer and drier interior highlands. The average annual rainfall in the latter is about 45 inches (1,140 millimetres) and, except for a short summer dry season, is evenly distributed throughout the year.

Plant and animal life. Only 3 percent of the country consists of natural forest vegetation. Reforestation programs have added eucalyptus trees to previously denuded hillsides and roadsides, though not on a scale sufficient to effectively counteract erosion. A lush, Mediterranean-type vegetation covers the shores of Lake Kivu, which stands in stark contrast to the papyrus swamps of the east and the thick bamboo forests of the Virungas. There, among the volcanoes, lives Rwanda's main tourist attraction, the mountain gorilla. For sheer diversity of animal life, however, no other region can match the resources of the Akagera National Park. Buffalo and zebras, antelope and warthogs, chimpanzees and lions, as well as many rare species—such as the giant pangolin, or anteater—are part of a fauna that also includes elephants, rhinoceroses, and hippopotamuses.

Settlement patterns. Despite a high population density, the dominant pattern is one of extreme dispersal. More than 90 percent of the population is rural and lives in nuclear family compounds scattered on hillsides. Kigali, the capital, was only a hamlet at the time of independence but has grown to become the largest city.

The people. *Ethnic composition.* As in Burundi, the major ethnic groups are Hutu and Tutsi, respectively accounting for almost 90 percent and about 10 percent of the total population. To these must be added the Twa hunter-gatherers, who constitute less than 1 percent of the population. Other minorities include a small group of Europeans, mostly missionaries and aid officials, a small number of Asian merchants, and a sprinkling of Swahili-speaking Africans from Tanzania and Zaire.

The clientage system

Though the Tutsi are generally taller and of lighter complexion than the Hutu, the physical stereotypes attributed to each group are greatly mitigated by intermarriage. Social differences between them, however, traditionally were profound, as shown by the system of patron-client ties (*buhake,* or "cattle contract") through which the Tutsi gained social, economic, and political ascendancy over the Hutu agriculturalists. During the revolution some 300,000 Tutsi were forced out of the country, thus reducing the former ruling aristocracy to an even smaller minority.

Linguistic composition. French and Rwanda (more properly, Kinyarwanda), a Bantu language closely related to Rundi, are the official languages of the country, but only a small fraction of the population speaks French. Swahili is widely spoken in the towns and is still the principal means of communication with Africans from neighbouring territories.

Religion. Nowhere in Africa has Christianity had a more decisive impact than in Rwanda. The Hutu revolution derived much of its egalitarian inspiration from the teachings of the European clergy, and Catholic seminaries served as recruiting grounds for Hutu leaders. Roman Catholicism claims the allegiance of about two-thirds of the population.

Demographic trends. Rwanda's rate of population increase is one of the highest in central Africa and is far above the productive capacity of the environment. Family-planning programs are virtually nonexistent. Approximately half of the population is under age 16; infant and child mortality rates are among the highest in Africa. Adding to the growing population pressure on the land, some 60,000 Hutu refugees from Burundi live in Rwanda, most of them having fled their homeland during the 1972 holocaust, when an estimated 100,000–150,000 Hutu perished in interethnic violence.

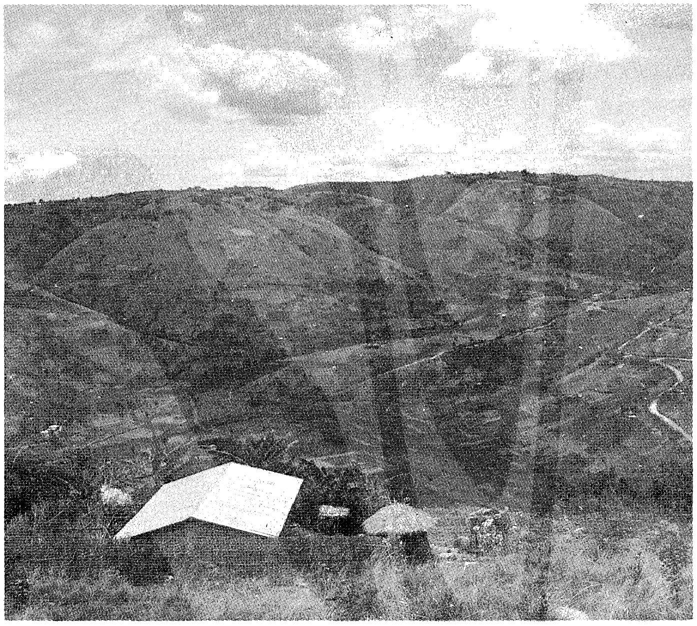
The economy. The country's economy is overwhelmingly agricultural, with coffee exports accounting for more than 70 percent of its foreign exchange and tea for more than 10 percent. An inadequate subsistence agriculture, however, is the dominant feature of the economy, with heavy infusions of foreign aid required to meet chronic food shortages. Rwanda is the largest per capita recipient of foreign aid in Africa. Mineral resources make up from 10 to 15 percent of total exports.

Resources. Mineral resources include, in addition to tin and tungsten (wolfram), tantalite, columbite, and beryl. Methane gas from Lake Kivu is used as a nitrogen fertilizer and is also converted into compressed fuel for trucks. What little gold Rwanda produces is smuggled out of the country. The Mukungwa hydroelectric power installation, the country's major source of electricity, meets only a portion of the country's energy needs, and much of the remainder must be imported from Zaire.

Major crops

Agriculture, forestry, and fishing. Intensive cultivation is practiced throughout the country, resulting in a large diversification of food crops. To the main indigenous crops—sorghum and eleusine—must be added corn (maize), potatoes, sweet potatoes, cassava (manioc), dry beans, and plantains. While eleusine is harvested in May and sorghum in July, bananas and plantains can be grown throughout the year. Arabica coffee (first introduced by European missionaries), tea, tobacco, and pyrethrum are the principal cash crops, with coffee constituting the prime export. Most of what is left of the small amount of natural forest is found on the slopes of the Virunga Mountains in the northwest. Fishing is widespread on Lake Kivu as well as in the smaller lakes of the interior, most notably Lake Muhazi and Lake Mugesera.

John Hatt/The Hutchison Library



Family settlements scattered on the deforested hillsides of Rwanda.

Industry. Aside from small-scale mining operations and limited consumer manufactures (such as textiles, soap, matches, cement, paint, some pharmaceuticals, and furniture), for the most part industrial activities involve the processing of coffee, tea, and other agricultural commodities. Most of the country's industries are located in Kigali.

Finance and trade. The sharp downward trend in the price of primary commodities since the 1970s, especially coffee and tea, has had a catastrophic effect on the country's balance of trade. It is estimated that the total civil service wage bill requires approximately half of the annual budget. Investment programs are almost entirely covered by external sources of financing. Efforts at promoting closer economic links, in the form of a customs union, between Rwanda and its neighbours through the Economic Community of Great Lakes Countries (CEPGL) have thus far met with little success.

Transportation. Rwanda claims one of the densest road networks on the continent, though only about 7 percent of it is paved. From Kigali to Ruhengeri and thence to Gisenyi, roads that were once impassable in the rainy season are now practicable throughout the year.

Administration and social conditions. *Government.* According to the 1978 constitution Rwanda is a "democratic, social and sovereign Republic whose . . . principle is 'Government of the people, by the people, for the people.' " This principle is strongly tempered, however, by the wide scope of authority vested in the presidency. The president of the republic, Major General Juvénal Habyarimana, combines the roles of head of state and head of government with that of president of the single ruling party, the National Revolutionary Movement for Development (Mouvement Revolutionnaire National pour le Development; MRND). He is elected for a five-year term by direct universal suffrage. As president of the MRND, he is the only candidate for the presidency of the republic. The legislative body is the unicameral National Development Council.

The presidency

Education. About half of the population is literate, and approximately two-thirds of the primary-school-age population is enrolled. Secondary schools have space for only 10 percent of the primary school graduates, however. The National University of Rwanda (1963) is in Butare. Ethnic quotas, reflecting the relative numerical strength of each ethnic group, have become the rule for regulating the flow of students at the secondary and university levels.

Health and welfare. Rwanda has one of the highest percentages of AIDS carriers in Africa; a quarter of a million cases were reported in 1985. Nutritional deficiencies present an even greater threat to the population, however, along with malaria and tuberculosis. Health facilities are still grossly inadequate, consisting for the most part of poorly equipped health centres and dispensaries. Welfare activities are primarily organized under the auspices of missionary societies.

Cultural life. *The arts.* Much of Rwanda's traditional cultural heritage revolves around dances, praise songs, and dynastic poems designed to enhance the legitimacy of kingship. Since the advent of the republic in 1962 another set of traditions has emerged, emphasizing a different cultural stream, identified with a Hutu heritage. Regional dances, including the celebrated hoe dance of the north, are given pride of place in the country's new cultural repertoire. Traditional crafts such as basketry, ceramics, and ironworks provide another element of continuity with the past.

Cultural institutions. Rwanda's National Ballet and the Impala Orchestra add considerable lustre to the country's cultural life, the former through a choreography leaning heavily on traditional folk dances and the latter through a distinctly modern musical repertoire. The Association des Écrivains du Rwanda (AER) keeps alive the best of Rwanda's literary traditions, while the quarterly review *Dialogue* provides a forum for a vigorous intellectual exchange on a wide range of social and cultural issues.

Press and broadcasting. The absence of stringent government censorship makes for a lively press. In addition to the weekly *Imvaho* (published in Kinyarwanda) and the monthly *La Relève,* at least two other publications deserve mention, the weekly *Kinyamateka,* published by Rwanda's Episcopal Council, and the monthly *Rafiki,* published by the Dominican Fathers.

For statistical data on the land and people of Rwanda, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Tutsi traditions trace the birth of the Rwanda kingdom to the miraculous feats of its founding hero, Gihanga, whose coming to Rwanda is said to coincide with the advent of civilization. A more sober appraisal, however, would emphasize a long process of Tutsi migrations from the north, culminating in the 16th century with the emergence of a small nuclear kingdom in the central region.

From 1894 to 1918, Rwanda, along with Burundi, was part of German East Africa. After Belgium became the administering authority under the mandates system of the League of Nations, Rwanda and Burundi formed a single administrative entity; they continued to be jointly administered as the Territory of Ruanda-Urundi until the end of the Belgian trusteeship in 1962. By then, however, the two states had evolved radically different political systems, with Rwanda having declared itself a republic in January 1961 and forced its monarch (mwami), Kigeri, into exile. Burundi, on the other hand, retained the formal trappings of a constitutional monarchy until 1966.

Origins of the revolution

The Rwanda revolution was rooted partly in a traditional system of stratification based on an all-embracing "premise of inequality" and partly in a colonial heritage that greatly increased the oppressiveness of the few over the many. Tutsi hegemony was unquestionably more burdensome under Belgian rule than at any time prior to European colonization. By the end of World War II, a growing number of colonial civil servants and missionaries had come to recognize the legitimacy of Hutu claims against the ruling Tutsi minority. The proclamation of the republic a year and a half before the country acceded to independence testifies to the substantial support extended by the trusteeship authorities to the revolution.

What began as a peasant revolt in November 1959 eventually transformed itself into an organized political movement aimed at the overthrow of the monarchy and the vesting of full political power in Hutu hands. Under the leadership of Grégoire Kayibanda, Rwanda's first president, the Party for Hutu Emancipation (Parti du Mouvement de l'Emancipation du Peuple Hutu; PARMEHUTU) emerged as the spearhead of the revolution. Communal elections were held in 1960, resulting in a massive transfer of power to Hutu elements at the local level. And in the wake of the coup in Gitarama in central Rwanda, which was carried off with the tacit approval of the Belgian authorities, in January 1961, an all-Hutu provisional government came into being. Thus by the time independence was proclaimed in July 1962, the revolution had already run its course, and the transition from a monarchical to a republican form of government had become reality.

With the elimination of Tutsi elements from the political arena, yet another reality came into view, in the form of regional competition among Hutu politicians. By the early 1970s the political landscape had become increasingly polarized by a north-south cleavage, reflecting the comparatively privileged position of elements from the central and southern regions within the party, the government, and the administration.

Regional tensions came to a head on July 5, 1973, when a group of army officers from the north overthrew the Kayibanda regime in a bloodless coup and installed Major General Juvénal Habyarimana. The seizure of power by the army meant that northerners would have a decisive voice in state affairs. As president of Rwanda's Second Republic, Habyarimana, a northerner, has given a distinctly regional coloration to the institutions of the state.

Creation of the Second Republic

With the coming to power of a government dominated by northern elements, north-south polarities eventually gave way to the emergence of subregional factions within the northern establishment. By 1980 the principal fault line of Rwanda politics reflected the continuing tug-of-war between Bashiru and Bagoyi elements, respectively identified with the Bushiru and Bugoyi subregions. By casting his lot with the Bashiru faction, Habyarimana became the target of considerable opposition from Bagoyi elements within the army. The arrest of scores of army officers following the disclosure of an abortive, Bagoyi-inspired coup in April 1980 was evidence of the continuing significance of subregional loyalties within the broader framework of Hutu identity.

Seen from a broader perspective, the political history of Rwanda remains profoundly influenced by its revolutionary heritage. The reconstruction of society on the basis of popular participation and social equality is the paramount objective of the Second Republic. The frequency of elections, the vitality of intraparty democracy, and the widespread commitment to self-help as a major ingredient of development all bear testimony to the country's revolutionary ideology. Whether the demons of regionalism and factionalism can be effectively exorcized to allow the Second Republic to further consolidate the gains of the revolution and to translate its promise into tangible benefits for the masses remains to be seen.

(Re.L.)

For later developments in the history of Rwanda, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

São Tomé and Príncipe

The Democratic Republic of São Tomé and Príncipe (República democrática de São Tomé e Príncipe) covers an area of only 386 square miles (1,001 square kilometres) and is thus one of the world's smallest countries. It consists of two small islands and a few islets in the Gulf of Guinea. São Tomé island is oval in shape, about 30 miles (49 kilometres) long, and 332 square miles in area. The capital of the country, São Tomé city, is situated in the northeastern part of the island. Príncipe lies about 93 miles to the northeast of its sister island, but it is smaller and has a more irregular outline, being about 13 miles long and 55 square miles in area. The country's closest neighbours are Gabon and Equatorial Guinea on the Atlantic coast of central Africa.

PHYSICAL AND HUMAN GEOGRAPHY

The land. In the south and west of both islands, high volcanic mountains fall precipitously to the sea, although neither island has witnessed any volcanic activity in recent centuries. The mountains descend gradually to small plains in the northeast. São Tomé Peak, the highest point on the main island, rises to 6,640 feet (2,024 metres) above sea level, while Príncipe Peak on the smaller island reaches 3,110 feet (948 metres). These mountainous areas are deeply dissected by the effects of stream erosion, and spectacular isolated volcanic plugs stand out as landmarks. Swift and rocky streams rush down to the coast in every direction.

Relief and climate

The climate is basically maritime and tropical, but because of the rough topography, there is a wide range of microclimates. The prevailing moist southwesterly winds are intercepted by the mountains, so that annual rainfall exceeds 275 inches (7,000 millimetres) in the southwest of São Tomé island, while the far northeast of the island receives less than 30 inches. The dry season lasts from June to September in the northeast but is scarcely discernible in the wetter regions. In the coastal areas the mean annual temperature is high, at about 81° F (27° C), and the average relative humidity is also high, at about 80 percent. Average temperatures decline sharply with altitude, and night temperatures fall to below 50° F (10° C) at about 2,300 feet. Above 3,300 feet fine misty rain falls almost continuously, and the nights are cold, although frost and snow are unknown.

The original vegetation of the islands was luxuriant tropical rain forest, with a gradual transition from lowland forest to mist forest. A little less than half the islands' area, mainly in the south and west, is still covered with rain forest. Much of this is secondary growth on abandoned plantation land. The flora and fauna include many rare and endemic species, reflecting the isolation and environmental diversity of the islands. Many of these plants, birds, reptiles, and small mammals are threatened by pressure on the remaining rain forest.

The population is concentrated in the drier and flatter areas of both islands. While a third of the population lives in São Tomé city and its outskirts, only about 5 percent

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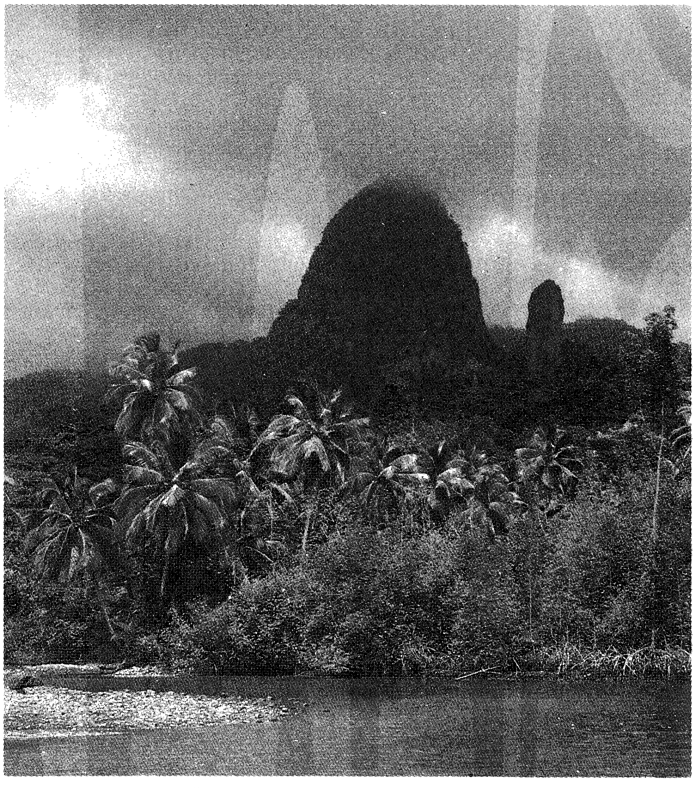


Two volcanic plugs, João Dias Pai (father) and João Dias Filho (son), standing in the mist along the Banzu River, on the west coast of Príncipe.

of the population inhabits the island of Príncipe. Villages are clustered along the roads and are often grouped around a church. Houses made of wooden planks and raised above the ground are typical of the local building methods although there are also many concrete structures in the Portuguese colonial style. Many people still live in barracklike accommodations on the plantations.

The people. The population consists mainly of Forros, descendants of immigrant Europeans and African slaves. They speak a creolized form of Portuguese; standard Portuguese is the official language and is understood by a majority of islanders. Cape Verdeans form the largest group of foreigners, perhaps amounting to about 10 percent of the total population; many have adopted São Tomense nationality. Portuguese, Angolans, and Mozambicans make up most of the rest of the foreign community. Like the Cape Verdeans, they are relatively well integrated with the other islanders, owing to a shared Portuguese cultural background. The Angolares, descended from slaves who survived a shipwreck in the 16th century, remained apart in the isolated southern zone of São Tomé island until the late 19th century, speaking a Bantu language. They have since spread throughout the country and have become largely assimilated. Almost the entire population belongs to the Roman Catholic church, although there are a few small Protestant congregations. Traditional African religions are also practiced.

Population growth is rapid, owing to both a relatively high birth rate—though it is below average for central Africa—and a low death rate. Nearly half the population is less than 15 years of age, assuring continued rapid growth into the future. Life expectancy is relatively high for an African country, at 63 years for men and 67 years for women.

The economy. Decades of colonial stagnation were followed by economic disruption after independence in 1975. Under the tutelage of the International Monetary Fund and the World Bank, São Tomé and Príncipe has gradually restored a functioning economy by devaluing its currency, restricting the budget deficit, privatizing formerly nationalized companies, attracting foreign investment, and removing price subsidies and controls. About four-fifths of the total land area of the two islands belongs to the state and is divided into 15 large plantation enterprises, several of which have been leased out to foreign management companies. High levels of unemployment coexist with a critical labour shortage on the plantations, where wages and working conditions are poor.

Economic resources

São Tomé is endowed with excellent conditions for tropical agriculture. The growing season is long, the volcanic soils are fertile, and there is no lack of water. There are some fine timber stands remaining, but the physical problems of removal in mountainous terrain and the pressing need for effective conservation limit long-term prospects. The small size of the country prevents the significant growth of extensive livestock raising, but conditions for poultry raising are quite favourable. There are numerous sites for small hydroelectric schemes but no large rivers for major installations. The country has no known minerals aside from some lime deposits exploited for local use. Fishing resources are limited by the narrow continental shelf. Manufacturing is hampered by the small size of the internal market, the limited energy resources, and the lack of skilled labour. Tourism is perhaps the best hope for economic diversification, but the long rainy season is a considerable obstacle to overcome.

The economy remains chronically dependent on plantation agriculture and especially on cocoa. About 40 percent of the total land area is under cultivation, with cacao trees covering a little less than two-thirds of the state-owned plantation acreage; coconut palms cover most of the remainder. Large areas of plantation land have been poorly maintained since independence, being harvested from time to time but not otherwise tended. The country has never been self-sufficient in staple foodstuffs, and a combination of official price-fixing for urban consumers and foreign gifts of food have undermined the production of food crops for the local market.

The nonagricultural sectors contribute little to the economy. The country is self-sufficient in fish, and trawlers from the European Communities pay small license fees for the right to fish in the country's national waters. The deep-sea tuna resources of the Gulf of Guinea and shellfish in coastal waters represent the best hopes for fishery exports. Manufacturing accounts for about 10 percent of the gross domestic product and consists mainly of small processing factories producing foodstuffs, beverages, soap products, and sawn wood for the internal market. Tourism is largely limited to the dry season and chiefly attracts people living in neighbouring parts of Africa.

São Tomé and Príncipe is reputed to be one of the highest recipients of foreign aid per capita in the world, but this has not prevented large budgetary and balance-of-payments deficits. There is only one bank active in the country, the state-owned National Bank of São Tomé and Príncipe, which controls foreign exchange dealings and acts as the central bank. The black market in foreign exchange has shrunk as the national currency, the dobra, has progressively been devalued. Cocoa accounts for almost all foreign exchange earnings. China takes cocoa under counter-trade barter agreements, but most of it is exported to Germany and The Netherlands. Portugal is the main source of imports, and Angola supplies petroleum products at concessionary rates.

Transportation assumes particular importance in this isolated microstate. There are no natural or artificial harbours, and large ships must anchor far out at sea and be unloaded by barge. Shipping links between the islands and with the outside world are erratic, and there are long delays in unloading cargo. The international airport near São Tomé city has been expanded and modernized, and in 1986 a national airline, Equatorial Airlines of São Tomé and Príncipe, was established as a joint venture between the government and foreign interests. It links the two main islands and flies to adjacent African countries. The telephone system and the road network are both fairly good by African standards.

Administration and social conditions. The political and judicial structures adopted at independence in 1975 were those of a single-party state, but the regime never formally proclaimed its adherence to Marxism-Leninism. Free elections for the legislative assembly and the presidency were established by the constitution of 1990. Close ties with eastern European countries have been replaced by warm relations with Portugal, France, and other Western countries. Angolan troops have been stationed in the country since 1978 to guarantee internal security.

National government

Educational and health levels are high by African standards. Some two-thirds of children receive basic education, and more than half the population is literate in Portuguese. Foreign aid workers make up for the shortage of teachers. There is no university, and most students go to Portugal for higher education.

Life expectancy is more than 60 years. There are numerous clinics, although drug supplies are short, local doctors emigrate, malaria is endemic, and there are occasional epidemics of cholera. However, there are no signs of extreme destitution.

Cultural life. This small country has a homogeneous culture, profoundly marked by centuries of Portuguese colonialism, although the government has been keen to stress the country's African heritage. Language, family structure, and religion are basically Portuguese. Many African elements have been adopted in the cooking, customs, beliefs, and dress of the common people, and the poorer classes only speak creole. The single newspaper appears erratically, but the government-run radio station is active, and there have been experiments with a television station.

For statistical data on the land and people of São Tomé and Príncipe, see the *Britannica World Data* section of the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Settlement by Portuguese

São Tomé and Príncipe were uninhabited when they were discovered, probably in 1470, by Portuguese navigators sailing around Africa. In the 1480s, the Portuguese sent out settlers (many of them Jewish "New Christians" expelled from Portugal by the Inquisition) and imported African slaves to grow sugar.

During the 16th century, São Tomé briefly became the world's largest producer of sugar, but the rise of Brazilian competition and the poor quality of São Tomé's badly dried product led to the decline of its sugar. This decline was accentuated by insecurity, as slaves escaped to the mountains and raided the plantations. Amador, the king of the fugitive slaves who nearly overran the whole island of São Tomé, is considered a national hero. Foreign pirates were another hazard, and the Dutch briefly captured São Tomé in 1641, only to be expelled three years later.

After the collapse of the sugar economy, the colony served as a bulking-up point for the Portuguese slave trade to Brazil; the cargoes of small slave ships were transferred to larger vessels for the Atlantic voyage, and provisions and water were obtained. The slave-worked plantations producing foodstuffs for the slave ships were mainly located on Príncipe, and the capital was moved to the port of Santo Antônio on Príncipe in 1753. In 1777–78 the Portuguese ceded the islands of Bioko and Annobón, on either side of São Tomé and Príncipe, to the Spaniards, who wished to develop their own African slave trade.

With the suppression of the Brazilian slave trade and the introduction of coffee cultivation in the 19th century, the economic centre of gravity swung back to São Tomé, and in 1852 São Tomé city once again became the capital. Cocoa replaced coffee as the main cash crop in the 1890s, and during the first two decades of the 20th century the colony was in some years the world's largest producer of cocoa. This led to the maximum expansion of plantations on the islands. Although slavery was legally abolished in 1875, Angolan slaves continued to make up the bulk of the labour force until the early 1910s. They were then replaced by forced or indentured labourers from other Portuguese colonies.

Cocoa production fell after World War I, and the islands progressively became a stagnant backwater, notorious for the brutality and corruption that reigned on the plantations belonging to absentee planters and corporations. Attempts to force the local Forros to work on the plantations led to the Batepá Massacre in 1953, seen as the beginning of the nationalist movement. The Committee for the Liberation of São Tomé and Príncipe was set up in exile in 1960, changing its name to the Movement for the Liberation of São Tomé and Príncipe (MLSTP) in 1972. However, it was never able to mount a guerrilla challenge to the Portuguese on the islands.

The nationalist movement

The Portuguese government that overthrew the Salazarist dictatorship in 1974 agreed to hand over power to the MLSTP in 1975, and many of the whites and wealthier Creoles then fled to Portugal, fearing a communist government. Independence was granted on July 12, 1975.

The MLSTP at first followed eastern European models of political and economic organization. Economic decline and popular dissatisfaction, however, led to a process of liberalization from 1985, culminating in the establishment of multiparty democracy in 1990. (W.G.C.-S.)

For later developments in the history of São Tomé and Príncipe, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Zaire

The Republic of Zaire (République du Zaïre) is centrally located on the African continent. It has a short 25-mile (40-kilometre) coastline on the Atlantic Ocean and is otherwise landlocked, being bordered on the west by the Angolan exclave of Cabinda and the People's Republic of the Congo; on the north by the Central African Republic and The Sudan; on the east by Uganda, Rwanda, Burundi, and Tanzania; on the southeast by Zambia; and on the southwest by Angola. It has an area of 905,446 square miles (2,345,095 square kilometres) and is the third largest country on the continent; only The Sudan and Algeria have larger areas. The capital, Kinshasa, is located on the Congo River at a distance of about 320 miles from its mouth and is the largest city in central Africa. It serves as the nation's administrative, economic, and cultural centre.

Formerly known as the Democratic Republic of the Congo, the country took the name of Zaire in 1971 in an attempt to return to the source of the nation's identity and authenticity. "Zaire" is a variation of traditional African names for big rivers and specifically the Congo River, whose basin lies almost entirely within the republic. The river was named during the colonial period for the kingdom of the Kongo people, who inhabit the area along the river's mouth on the Atlantic Ocean.

Zaire is a country rich in economic resources. Its minerals include vast deposits of industrial diamonds, cobalt, and copper; its forest reserves are possibly the largest in Africa; and its hydroelectric potential comprises half that of the African continent.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* The country's major relief features include the coastal region, two major basins or depressions, high plateaus, and three mountain ranges. The narrow coastal region is composed of a fairly low plain that runs inland from the Atlantic Ocean to the Cristal Mountains, where a high escarpment rises above the plains.

Most of the country is composed of the central (or Congo) basin, topographically a vast rolling plain with an average elevation of about 1,700 feet above sea level. Its lowest point of 1,109 feet (338 metres) occurs at Lake Mai-Ndombe (formerly Lake Léopold II), and the highest point of 2,296 feet (700 metres) occurs in the hill country of Mobayi-Mbongo and Zongo in the north. This basin may once have been an inland sea whose only remaining vestiges are now Lakes Tumba and Mai-Ndombe in the west-central part of the basin.

A high longitudinal basin—the western arm of the East African Rift System—forms the country's eastern border. Along its Zaire section, the depression contains Lakes Albert (in Zaire known as Lake Mobutu Sese Seko), Edward, Kivu, Tanganyika, and Mweru.

The high plateaus

The high plateaus border almost every side of the central basin. In the north the basin is protected by the Ubangi-Uele plateaus forming the divide between the drainage basins of the Nile and Congo rivers. Rising to between 3,000 and 4,000 feet, the plateaus also separate the central basin from the vast plains of the Lake Chad system. In the south the plateaus begin at the lower terraces of the Lulua and Lunda river valleys and rise gradually toward the east. In the southeast the ridges of the plateaus of Shaba tower over the entire area; they include Kundelungu at 5,250 feet (1,600 metres), Mitumba at 4,921 feet (1,500 metres), and Hakansson at 3,609 feet (1,100 metres). The Shaba plateaus extend as far north as the Lukuga River and contain the Manika Plateau, the Kibara and the Bia mountains, and the high plains of Marungu.

The northern escarpment of the Angola Plateau rises in the southwest. In the west there is a coastal plateau zone that includes the hill country of Mayumbe and the Cristal Mountains. Mount Ula at 3,446 feet (1,050 metres) is the highest point of the Cristal Mountains.

The eastern part of the country is the highest and most rugged. It contains striking chains of mountains that are part of the East African Rift System. The Mitumba Mountains stretch along the Western Rift Valley, rising to an elevation of 9,800 feet above sea level. The snow-covered peaks of the Ruwenzori Range between Lakes Albert and Edward lie astride the Uganda border and contain Zaire's highest elevation of 16,795 feet (5,119 metres) at Margherita Peak. The Virunga Mountains, to the north of Lake Kivu, form a volcanic range that stretches across the Rift Valley.

The Congo River basin

Drainage and soils. The Congo (Zaire) River, including its 1,335,000-square-mile basin, constitutes the main system of drainage in the country. The river rises in the high Shaba plateaus and flows north and then south to cross the equator twice in a great arc. The lower river flows southwestward to empty into the Atlantic Ocean below Matadi. Along its course the Congo passes through alluvial lands and swamps and is fed by the waters of many lakes and tributaries. The most important lakes are Mai-Ndombe and Tumba; the major tributaries are the Lomami, Aruwimi, and Ubangi rivers and those of the great

Kasai River system. There is also a link to the Western Rift Valley via the Lukuga River.

There are two types of soils: those of the equatorial areas and those of the drier savanna (grassland) regions. The equatorial soils occur in the warm, humid lowlands of the central basin, which receive abundant rainfall throughout the year and are covered mainly with thick forests. This soil is almost fixed in place because of the lack of erosive forces in the forests. In the shore areas, however, swamp vegetation has built up a remarkably thick soil that is constantly nourished by humus, the organic material resulting from the decomposition of plant or animal matter.

Although in the savanna regions the soils are constantly endangered by erosion, the river valleys contain rich and fertile alluvial soils. Special note should be made of the highlands of eastern Zaire in the Great Lakes region, which are partly covered with volcanic lava that has been transformed into exceptionally rich soil. This is the most productive agricultural area of the country.

Climate. The major part of Zaire lies within the inner humid tropical, or equatorial, climatic region extending five degrees north and south of the equator. Southern Zaire and the extreme north have somewhat drier subequatorial climates.

The seasonally migratory intertropical convergence zone (ITCZ) is a major weather feature of Zaire's climate. Along this zone the trade winds originating in the Northern and Southern hemispheres meet, forcing unstable tropical air aloft. The resulting cooling and condensation of the uplifted air produces prolonged and heavy rainfall. In July and August this zone of maximum rainfall occurs in the north. It passes into central Zaire in September and October. From November to February the southern parts of the country receive their maximum precipitation. Moving northward, the ITCZ crosses central Zaire again in March and April, so that this zone has two rainfall maxima. Only the extreme eastern highlands lie outside the path of the ITCZ and are subject to the influence of the southeastern trade winds alone. Altitude and proximity to the Atlantic Ocean and its maritime influences are additional factors of climatic differentiation.

The four major climatic regions

There are four major climatic regions in the country. In the area of the equatorial climate, temperatures are hot, and the average monthly temperature rarely drops below 75° F (24° C). Humidity is high, and it rains almost throughout the year. Annual precipitation at Eala averages 71 inches (1,800 millimetres). The tropical or subequatorial climate occurs to the north and south of the equatorial region. It is marked by distinct dry and rainy seasons. The dry season may last from four to seven months (usually April to October) of the year, depending largely on distance from the equator. At Kananga about 62 inches of rain falls annually. In some areas and at indefinite periods, a short dry season of several weeks' duration may occur during the rainy season.

The Atlantic climate is limited to the west coast. It is marked by the modifying influences of low altitude and the cold Benguela Current. At Banana the average annual temperature is 77° F (25° C), and rainfall averages about 30 inches yearly. The mountain climate occurs in the eastern high plateaus and mountains. At Bukavu the average annual temperature is 66° F (19° C), and annual precipitation measures about 52 inches.

Plant and animal life. Plant life is profuse and follows climatic patterns. In the centre of the Congo basin is an intricate forest system, commonly known as the equatorial rain forest. There trees reach 130 to 160 feet in height, and many plant varieties and species can be found in a small area. In the tropical climate zone, grassland and woodland are characteristic, while in the west the coastal swamps and the mouth of the Congo are dominated by stands of mangrove. The eastern plateaus are covered by grasslands. Mountain forest, bamboo thickets, and Afro-Alpine vegetation occur on the highest mountains.

The central basin is a vast reservoir of trees and plants that are native to the area. Among these, the mahogany, ebony, limba, wenge, agba, iroko, and sapele are sources of timber. Fibrous plants include raffia and sisal. There are also plants used in traditional medicine, including cinchona (the source of quinine) and rauwolfia (an emetic and antihypertensive), as well as copal, rubber, and palm trees. Many types of edible mushrooms grow wild; other wild edible vegetables grow in the forests, grasslands, and swamps. Eucalyptus trees have been imported and form important stands in the highlands; they are used for construction timber and poles.

Animal life is also rich and diversified. Chimpanzees are found mostly in the equatorial forest, and gorillas occur in the eastern mountains around Lake Kivu. Elephants and various species of monkey and baboon are found in the forest and the savanna woodland. The short elephants are, however, exclusively forest-bound.

In the primary forests of Uele, Aruwimi, and Ituri are the okapi, the giant wild boar, and the short antelope. The lion and leopard inhabit the grasslands, and the jackal, hyena, cheetah, wildcat, wild dog, buffalo, antelope, wild hog, and black and white rhinoceroses are found in the grasslands and savanna woods. Giraffes mainly inhabit the northeastern grasslands.

Hippopotamuses and crocodiles are common in the rivers and lakes, and whales, dolphins, and lungfishes are found near the coast. Zairian rivers, lakes, and swamps are well stocked with a variety of fish, such as the capitaine from the Congo River and catfish, electric fish, eels, cichlids, and many others. There is also a good supply of jellyfish in Lake Tanganyika. Reptiles are common and include various snakes, such as pythons, vipers, and tree cobras, as well as lizards, chameleons, salamanders, frogs, and turtles.

Birdlife includes the pelicans, parrots, and many species of sunbird, pigeon, duck, goose, eagle, vulture, cuckoo, owl, crane, stork, and swallow. The insects are innumerable. There are hundreds of species of butterfly; in the savanna woodlands the butterflies have their special season at the beginning of the rains, when they can be seen flying in great numbers, filling the sky and wandering over the blooming trees. There are also bees of all types, different species of grasshopper, and caterpillars, praying mantises, beetles, dragonflies, scorpions, mosquitoes, tsetse flies, ants, termites, spiders, centipedes, and millipedes.

National parks

Much of the animal life has diminished as the result of hunting, which is now strictly regulated. Several national parks have been created, most in the eastern highlands. They include Garamba, near the Sudan border; Virunga, north of Lake Edward in the Virunga Mountains; Maiko, west of Lake Edward; Kahuzi-Biega, north of Bukavu; Upemba, north of the Manika Plateau; and Kundelungu, near the Zambian border northeast of Lubumbashi.

Settlement patterns. The traditionally inhabited regions are the forests, savanna woodlands, and grasslands. People have worked these areas and have become specialized to their environment. The individuals who live in the forests, such as the Bambuti (Pygmies) of the Ituri Forest, have specialized mostly in hunting and fishing, while agriculture has remained secondary or is nonexistent.

In the savanna woodlands the inhabitants combine cultivation with hunting and fishing. In some areas across the southern half of the country, the people are engaged in the raising of small livestock and poultry and traditionally mine copper, iron ore, and other minerals. In the grasslands inhabitants confine themselves almost solely to agriculture. In the eastern grasslands, however, agriculture is combined with the raising of large livestock.

Rural settlement

A large percentage of the Zairian population is rural, and most of the people live in scattered villages. The style of housing varies regionally, as does the general size of the villages. A village with 10 to 25 housing units is generally considered small, while one with 150 to 200 is considered large. The savanna woodlands of the south-central regions, and to some extent the coastal region, are the most populous areas, with the largest villages having 300 to 500 people. The eastern grassland areas have isolated farms and hamlets.

Such trade and administrative centres as Banana, Vivi, and Boma were established with the arrival of the Europeans in the 16th century. Most towns, however, are of more recent origin. Kinshasa, until 1966 called Léopoldville, is the seat of national political, administrative, and judiciary institutions and is also an important commercial and industrial centre. It is the creator and propagator of fashions and of many traits of Zairian cultural life. The growth of Kinshasa typifies that of many of the country's cities. In 1889 it had a population of 5,000; in 1925, when it was recognized as a *ville* (urban centre), it had grown to 28,000. The city jumped to a population of 250,000 in 1950, 1,500,000 in 1971, and about 3,500,000 in 1989—an increase by a factor of 1,000 in a century.

There are nine other major cities; all are administrative or commercial centres with the exception of Likasi, which is mainly an industrial and mining town. Kananga is the capital of Kasai-Occidental (Western Kasai) region. Lubumbashi (formerly Élisabethville), the administrative headquarters of Shaba, is the heavily industrialized capital of the nation's copper-mining activities. Mbuji-Mayi is the capital of Kasai-Oriental (Eastern Kasai) region and the country's diamond centre. Kisangani (formerly Stanleyville), the terminal point of navigation on the Congo River from Kinshasa, is the capital of Haut-Zaïre. Bukavu, the headquarters of Sud-Kivu region, is a major tourist centre; Kikwit, the former capital of Bandundu region, is the terminal port on the Kwilu River; and Matadi, the capital of Bas-Zaïre, is the country's main port. Mbandaka is a river port and the capital of Équateur region.

All of these towns were developed during the colonial period, when there were separate sectors for the Europeans and the Africans. The European neighbourhoods were characterized by big houses with large yards, wide and paved streets, and adequate electricity. The crowded African areas had smaller houses and yards and poor, if any, electric supply. These characteristics still hold true, although the traditional European sectors include Africans of the upper income group.

The people. *Ethnic composition.* It is common for the modern African to prefer an identification as simply an African or as a citizen of a particular country. It is possible, however, to distinguish ethnic, linguistic, or cultural groups among the Zairian population.

The Bantu peoples constitute a large majority of the country's population and occupy more than two-thirds of the national territory. They entered the region of modern Zaire during the 10th to the 14th century from the west and north and established kingdoms that were flourishing at the time of European penetration after the 16th century. The major kingdoms were those of the Kongo, Teke (Bateke), Luba, Pende, Yaka, Lunda, Songe, Tetela, and Kuba peoples. Major cultural clusters today include the Mongo (in the centre), Kongo (west), Luba (south-central), Lunda (south), Bemba (southeast), and Kasai (southwest). Bantu tribes in the north and northeast include Ngala, Buja, Bira, Kuumu, and Lega (Rega).

The Pygmies are considered the earliest inhabitants of the Congo basin, having arrived possibly during the Upper Paleolithic Period. The remaining Pygmies—the Bambuti, Twa, and Babinga—inhabit the forests of Kibali and Ituri and the regions of Lakes Kivu and Tanganyika and the Lualaba, Tshuapa, Sankuru, and Ubangi rivers.

There are other small non-Bantu African populations. The Sudanese groups who settled in the north include the Zande (Azande), Mangbetu, Banda, and Barambu (Abarambo). The Nilotic peoples live in the northeast; they include the Alur, Kakwa, Bari, Lugbara, and Logo. Hamitic peoples from North Africa and Rwanda are few; they include the Tutsi, who live in the lake region.

The permanent European and Asian population comprises about half of the country's aliens. Most of them came to Zaire for temporary employment. Much of the remaining alien population is composed of Africans of non-Zairian nationality.

Linguistic composition. More than 200 languages and dialects are spoken in Zaire. Communication between groups has been facilitated by four national languages: Swahili, Tshiluba (Kiluba), Lingala, and Kikongo. French is the only official language and the language of instruction, business, adminstration, and international communications. The four national languages are used in local trading and radio broadcasting. Lingala is growing rapidly; it is the official language of the military and is widely spoken in Kinshasa, where it is used in popular music.

Religion. The traditional religious beliefs in a supreme being, the power of the ancestors, spirits of nature, and the efficacy of magic were torn apart or greatly disturbed with the introduction of Christianity. There is a sizable Christian population, including the local sect of the Church of Jesus Christ on Earth by the Prophet Simon Kimbangu (Kimbanguism). The rest of the African population continues to follow traditional beliefs or professes no religion. The foreign community includes a small Jewish population and some Hindus and Muslims.

The economy. Located in the centre of the African continent, the Republic of Zaire—with its great size, population, and rich potential—is called upon to play an important regional and international economic role.

Mineral resources

Resources. The country's main economic resource is its mineral deposits. The abundance of minerals in the Shaba region (Swahili *shaba,* "copper") was a source of the desire of European powers to control the area. Minerals of Shaba include copper, cobalt, zinc, cassiterite (the chief source of metallic tin), manganese, coal, silver, cadmium, germanium (a brittle element used as a semiconductor), gold, palladium (a metallic element used as a catalyst and in alloys), and platinum.

The region west of Lake Kivu contains cassiterite, columbotantalite, wolframite (a source of tungsten), beryl, gold, and monazite (a phosphate of the cerium metals and thorium). Lake Kivu has a vast reserve of methane, carbonic, and nitrogen natural gases. There are deposits of iron ore in south-central Zaire. Industrial diamonds are found in the central regions, and gem-quality diamonds occur in the south-central part of the country.

There are gold, coal, and iron-ore deposits in northeastern Zaire, and there are prospective deposits of gold, monazite, and diamonds in the northwestern regions. The diamond deposits in the western region are insignificant for industrial exploitation. Coastal Zaire contains bauxite, gold, and offshore deposits of petroleum. The limestone deposits that occur throughout the country are considered to be among the richest in Africa.

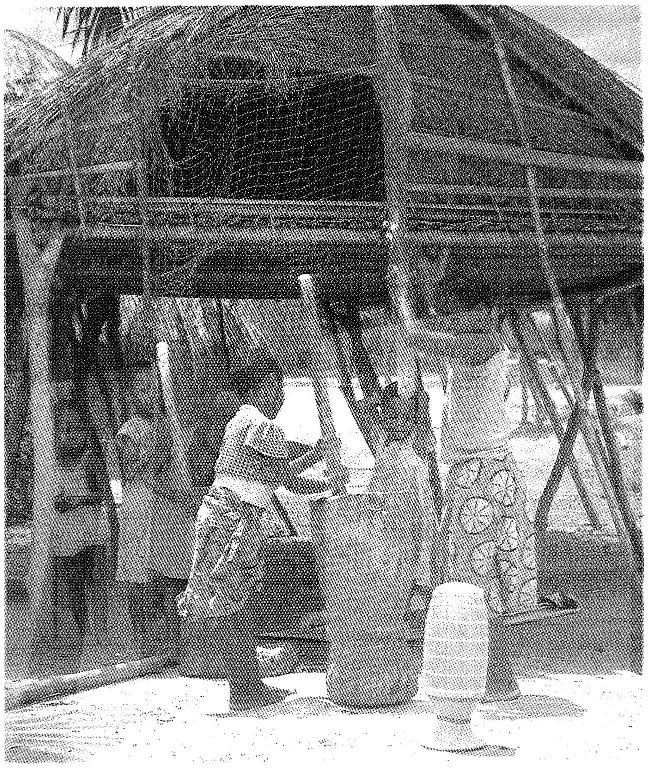
Zaire's forest reserves cover more than half of the country and are considered to be the largest in Africa. The wide variety of wild game supplements the local diet and contributes to a certain extent to local commerce. Ivory is exploited in the north, and the rivers, lakes, swamps, and ocean contain a vast reserve of fish.

The country's hydroelectric resources have an estimated potential of 13 percent of the world's capacity and 50 percent of Africa's potential capacity. This tremendous potential is derived from the many rapids along the rivers of the Congo system. Thermal energy can be derived from the forests and coal and petroleum deposits, as well as the uranium deposits in Shaba region.

Agriculture, forestry, and fishing. Domestic agriculture is the main source of food supply and cash income for the majority of the population. Although the country is rich in agricultural potential, the deterioration of the transportation network and agricultural services since independence have led to a regrowth of subsistence agriculture and a collapse of market production. Foodstuffs such as cereals and fish must be imported in increasing amounts. Coffee is the chief agricultural export; palm oil, rubber, and cotton, once mainstays of the export economy, have become almost negligible.

In the humid equatorial region, cassava (manioc) and rice are the basic food crops. Peanuts (groundnuts), oil palms, and fruit trees are also important, while robusta coffee is the main cash crop. In the eastern highlands, yams, beans, and sweet potatoes are used as food crops, while arabica coffee and tea are export commodities. On the southern plateaus, corn (maize) is of major importance for the urban populations of Shaba. Vegetable growing is widespread throughout Zaire.

Livestock and poultry are kept in every province. Cattle are raised mainly in the eastern and southern regions. Pigs are kept in the west and sheep in the eastern highlands. Other farm animals include chickens, geese, pigeons, and rabbits. Commercial meat production is limited, however, and the country depends upon imports to fulfill most of its requirements.



Lunda children pounding cassava into flour, southwestern Zaire.
John Anthony/Bruce Coleman Ltd.

A small part of the yearly production of timber is exported for veneering or plywood; most, however, is used locally for fuel. There is some commercial freshwater and ocean fishing. Local hunting and fishing for private consumption are not ordinarily reported in official statistics and are difficult to measure.

Industry. Mining produces more than half of the national budget and more than 80 percent of total exports. Zaire is a leading producer of industrial diamonds, accounting for about one-third of the world's total production. It also produces about one-half of the world's cobalt. It is a major producer of copper and tin. Coal production is low, however, because of mining difficulties and the increasing production of hydroelectricity. Other minerals mined include cadmium, silver, manganese, gold, wolframite, columbotantalite, beryl, and monazite. The most important mining company is the state-owned Générale des Carrières et des Mines (Gécamines).

Mining and quarrying

Manufacturing industries can be classified into two main categories. Consumption industries produce processed foods, beverages, cigarettes, cloth, printed material, hosiery, shoes and leather, metallic fabrics, and such chemical products as soap, paints, rubber, and plastics. The supply and equipment industries include spinning and weaving plants, chemical factories, and facilities to produce machinery, transport materials, nonmetallic minerals, and wood products. A petroleum refinery near Moanda was established in 1968; the production capacity is 750,000 tons annually.

The heaviest concentration of hydroelectric consumption is in the mining areas and in Kinshasa. The hydroelectric dam completed in 1972 on the lower Congo River at Inga Falls began initially to supply 300,000 kilowatts of electricity. After the completion of the second stage in 1982, Inga's capacity rose to 2,300,000 kilowatts. The dam has a theoretical potential estimated at 30,000,000 kilowatts. Zaire exports electricity to Zambia, Burundi, Congo, and Angola. There are thermal power plants in almost every major city that cannot be served by hydroelectric stations.

Finance and trade. The national central bank, the Bank of Zaire, is located in Kinshasa, as are numerous commercial, savings, and development banks. Most of these banks maintain branch offices in the regional capitals and major cities. There are also mortgage and credit banking institutions. Totally foreign-owned banks include U.S., British, and French institutions as well as the International Bank for Africa in Zaire.

Zaire has a favourable balance of trade. Mineral products account for most of total exports. The second most valuable exports are agricultural products; exported manufactures are of limited value and volume. Imports consist of consumer goods, machinery (largely mining and transport equipment), construction materials, leather and textiles, fuel, chemical products, metal products, and increasing amounts of foodstuffs.

Transportation. The organization of the transportation network is of the most crucial importance to Zaire, a country of continental dimensions and rich economic resources. The country's generally poor transportation infrastructure is a major factor in the underdevelopment and stagnation of the economy. The Congo River, the spinal cord of the country, and its tributaries serve as the main transport arteries. The rivers are supplemented by rail, road, and air services.

Navigation is possible throughout the year on sections of the Congo River. It is navigable from Banana to Matadi, Kinshasa to Kisangani, from Obundu to Kindu, and from Kongolo to Bukama for a total of 1,428 miles. Its tributaries add at least another 8,750 miles of navigable rivers. The main port for maritime shipping is Matadi near the mouth of the Congo River.

Railways

The agricultural region of Mayumbe is served by the Boma-Tshela railway. Other lines connect the Uele with the Itimbiri River and Lake Tanganyika with the Lualaba River, and two railways serve the rich southern regions.

There are four major routes that combine water and rail transport. The only such route to lie wholly within Zaire runs by rail from Shaba to Ilebo, by boat along the Kasai and Congo rivers to Kinshasa, and by rail to Matadi. The international routes run across Lake Tanganyika and Tanzania to the Indian Ocean port of Dar es Salaam; through Zambia, Zimbabwe, and Mozambique to Beira, also on the Indian Ocean; and through Angola to the Atlantic Ocean port of Lobito. The Angolan route and the system to Beira have been unusable for years, however, because of ongoing civil wars in Angola and Mozambique. The heavy traffic that would normally follow these routes is sent via Zimbabwe to ports in South Africa. In an effort to lessen its dependence upon its neighbours, the government plans a railway to connect Shaba directly with its Atlantic Ocean ports by linking Ilebo with Kinshasa.

Administration and social conditions. *Government.* The constitution of 1978 was Zaire's fifth since independence in 1960; trends apparent in these successive documents include the institution of an increasingly centralized presidential form of government, the exclusion of political activity not sanctioned by the Popular Movement of the Revolution (Mouvement Populaire de la Révolution; MPR), and the progressive circumscription of the prerogatives of the national legislative body by the Political Bureau of the MPR.

The MPR

The MPR is the sole legal political party. It is presided over by the president, who is assisted by the Political Bureau, composed only of individuals appointed by the president. The routine of party life on the national level is carried on by the Central Committee, whose members are appointed by the Political Bureau. The MPR has branches at every administrative level throughout the country. Party ideology is that of an anti-tribalist nationalism; the ultimate goal of building a strong nation-state is prescribed in the basic party document, the Nsele Manifesto (Le Manifeste de la N'sele).

Citizens of Zaire become members of the MPR at birth and participate in political activities by following the party teachings, voting, attending political rallies, and mobilizing on all levels for political and economic development.

According to the provisions of the 1978 constitution, the president of the republic is elected to a seven-year term (not to exceed two terms) consequent upon his election to the leadership of the MPR; the general electorate approves this choice. The president is the head of state, of the National Executive Council, and of the National Legislative

Council (the parliament) and is commander in chief of the armed forces.

The country is divided into eight *régions* and the *ville neutre* (equivalent to a federal district) of Kinshasa. Each region is divided into *sous-régions,* each *sous-région* into *zones,* each *zone* into *collectivités locales,* and each *collectivité locale* into *localités.* The *régions, sous-régions,* and *zones* are presided over by commissioners and the *collectivités locales* and the *localités* by chiefs. All are appointed to five-year terms, with the exception of the village chief, who has a completely traditional role and is elected by the villagers or may inherit the position.

Cities are divided into *zones* and *zones* into *quartiers,* headed, respectively, by *commissaires sous-régionaux,* or mayors, *commissaires des zones,* and *chefs de quartiers.* The mayor of Kinshasa is called *commissaire urbain.*

Justice. The judicial system is headed by the Supreme Court, which is located in Kinshasa. There are nine Courts of Appeal, and a tribunal of first instance is located in most subregional capitals.

The customary judicial system in the villages is not completely abolished, and villagers may settle some cases according to traditional procedures. Zaire does not apply the jury system; the judges, after deliberation, have the right to settle cases. All courts are presided over by a lawyer or person of legal training. The defendant may defend himself or choose his own lawyer or defender.

Education. Since independence public authorities have recognized the value of education and have given it greater attention. The budget for the Ministry of Education has always been ahead of that of the other sectors of national life. Primary education is compulsory, although it is difficult for a developing nation to meet this pledge because of the lack of facilities and an inadequate number of teachers.

The National University of Zaire

In 1971 the universities of Kinshasa, Kisangani, and Lubumbashi were merged into the National University of Zaire, which used the three campuses to house different departments and fields of study. This scheme was terminated in 1981, when the three former universities were reconstituted as separate, autonomous institutions by the Central Committee of the MPR. There are university institutes in the three university towns and at Buvaku and two arts academies in Kinshasa.

Health and welfare. In general, individuals build their own houses according to their needs and means. The government has established a department that builds and rents houses and also sells condominiums, especially in urban areas. In the cities, real-estate agencies and individuals build houses and apartments and make them available for rent. The government has been trying to improve housing conditions by the allocation of funds for the construction and improvement of houses.

In 1960 Zaire inherited a difficult medical situation, for there were no national doctors. The colonial administration had trained some highly qualified medical technicians and nurses while confining medical practice to European doctors and missionaries. By the late 1970s, however, most of the doctors were Zairian. For the country's first decade, experienced medical assistants, technicians, and nurses filled the vacuum left by the shortage of doctors. By the end of its third decade there was one doctor for every 14,000 persons.

With the means at its disposal and the help of such international organizations as the World Health Organization and the United Nations Children's Fund, the government has been waging a battle against the most critical and widespread diseases—tuberculosis, sleeping sickness, leprosy, and polio. Smallpox was eradicated in 1972. Special centres and programs have been established in the cities and rural areas to deal with maternity and child care, sanitary education, sanitary improvement of the environment, and preventive and curative medicine. A hospital boat equipped with doctors and modern facilities serves the people along the Congo River and its tributaries.

Zaire has an extensive medical care program for all workers. Each company is required to provide medical assistance for its employees, and public servants receive free medical treatment from state hospitals and dispensaries. Medical expenses in public hospitals are nominally low and accessible to people with minimum incomes, but shortages of equipment, supplies, and personnel make services difficult to obtain. There are a number of hospitals and dispensaries, including specialized hospitals for leprosy, neuropsychiatry, and tuberculosis.

Cultural life. Zaire's many ethnic groups and regions have developed a mosaic of traditional arts, such as painting, sculpture, music, and dance. There has been a tendency to classify sculpture and carving according to the styles of the areas from which they originate. The South-West Group is represented by the Kongo people and is known for stone and nail-studded statues; the Yaka, also of this group but from the region of the Kwango and Kwilu rivers, produce masks and figurines. The Kuba art form, the Southern Central Group, deals with human themes; there is a tendency to create statues representative of the Kuba royal court. The South-East Group is composed mainly of Luba art, which reflects the strong influence of women in the society through a multitude of statuettes relating to motherhood. North of the Luba region is the area where the Lega masks and ivories are produced. The North Group includes Zande and Mangbetu art; Zande art is characterized by cult statuettes, spear or bow shafts, and anthropomorphic pottery, while Mangbetu art is characterized by stylized elongated heads.

Traditional art styles

Painting, sculpture, and carving are still practiced. The cities, especially Kinshasa, have become the greatest creators, propagators, and promoters of national cultural life and arts. The Academy of Fine Arts in Kinshasa offers training programs in painting, sculpture, carving, architecture, and ceramics. The National Institute of the Arts offers training in classical and some traditional music and drama. The writing of poetry, plays, and novels has been developing rapidly; they are usually written in Lingala and French, but other local languages are also used. The development of scientific literature is supported by the universities, various scientific organizations, and the government. Several publishing houses have been established throughout the country.

Surviving national folk traditions are evident in pottery and the weaving of raffia, in ceremonial dress or costumes, in dancing styles, and in songs.

Sports include soccer, swimming, boxing, basketball, and riverboat racing. Zaire's unique popular music results from a mixture of traditional rhythms and instruments borrowed from other cultures, civilizations, and continents. This music, popular all over Africa, has given birth to the great variety of specific dance steps and styles known as the Zairian dance.

There are museums and public libraries in most large cities, with national museums in Kananga, Mbandaka, and Lubumbashi. The capital city houses the national archives and the National Theatrical Troupe. There are libraries at all three universities. (N.Pa./B.M.W.)

For statistical data on the land and people of Zaire, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

What began as a king's private domain (the Congo Free State), evolved into a colony (the Belgian Congo), and came to be known at the time of independence in 1960 as the Republic of the Congo (later the Democratic Republic of the Congo) and in 1971 as Zaire, is the product of a complex concatenation of historical forces. Some are traceable to the precolonial past, others to the legacy of colonial rule, others still to the political convulsions that followed in the wake of independence. All, in one way or another, have left their imprint on Zairian societies.

Precolonial perspectives. Before experiencing the radical social transformations of the colonial era, Zairian societies had already suffered major disruptions. From the 15th to the 17th century several important state systems came into existence in the savanna region in the southern half of the area. The most important were the Kongo kingdom in the west and the Luba and Lunda empires in the east. All three developed fairly elaborate political structures, buttressed by the symbolic force of kingship as well as by military force. Typically, power emanated

Kongo, Luba, and Lunda kingdoms

from the capital city to the outlying areas through the intermediation of appointed chiefs or local clan heads. Competition for the kingship often led to civil strife, however, and with the development of slave-trading activities a new source of instability was injected into regional politics. The history of the Kongo peoples in the 16th century is largely the story of how the Atlantic slave trade created powerful vested interests among provincial chiefs, in time greatly lessening the capacity of the kingdom to resist the encroachments of its neighbours. Thus, in the late 16th century, the kingdom had all but succumbed to the attacks of the Jaga, a group of warriors from the east. Two centuries later the Lunda and Luba peoples underwent a similar process of internal fragmentation followed by attacks from various interlopers, including Arabs and mestizos, eager to control the trade in slaves and ivory. On the eve of the European conquest their political institutions were both fractious and oppressive.

The radically different ecological conditions prevailing in the tropical rain forest raised formidable obstacles in the way of state-building. Small-scale segmentary societies, organized into village communities, were the rule. Corporate groups that combined social and economic functions among small numbers of related and unrelated people formed the dominant mode of organization. Among such corporate groups, exchange took place through trading activities and reciprocal gift-giving. Social interactions in time produced a measure of cultural homogeneity among otherwise distinctive communities, as among Bantu and Pygmy. Bantu communities absorbed and intermarried with their Pygmy clients, who brought their skills and crafts into the culture. The element of continuity discernible in the persistence of house and village organization stands in sharp contrast with the more centralized state structures characteristic of the savanna kingdoms. Nonetheless, on the eve of the Belgian conquest most Zairian societies had reached a degree of internal decomposition that greatly lessened their capacity to resist a full-scale invasion.

Resistance to outside forces in the savanna region was hampered by the devastating raids and civil wars that followed in the wake of the slave trade, by the improved capacity of Africans to destroy each other through the use of firearms, and ultimately by the all-too-familiar divisions between collaborators and resisters. The relative ease with which Zairian societies yielded to the European conquest bears testimony to the profound internal dislocations most of them had experienced in the course of previous centuries.

The Congo Free State. The catalyst for organizing the conquest of the huge domain that was to become his personal fief was King Leopold II of the Belgians. His thinly veiled colonial ambitions paved the way for the Berlin West Africa Conference (1884–85); thus armed with a mandate of international legitimacy, and under the cover of the African International Association, Leopold created one of the most coercive instruments of colonial hegemony. What became known as the Congo Free State (1885–1908) is most readily associated with the extraordinary hardships visited upon the Congolese masses in the name of the Leopoldian "civilizing mission." "Without the railroad," said Leopold's agent, the British explorer and journalist Henry Morton Stanley, "the Congo is not worth a penny." Without recourse to forced labour, however, the railroad could not be built, and the huge concessions made to private companies would not become profitable; nor could African resistance in the east be overcome without a massive recruitment of indigenous troops. The cruel logic of the revenue imperative left Leopold with no other option than to transform his nascent administrative system into an infernal machine designed to extract a maximum output of labour and natural resources from the land.

Necessity of the railroad

Not until 1908 did the Belgian Parliament vote in favour of annexation as the most sensible response to the flood of international criticism generated by the Leopoldian system. The impact of the Free State on the African populations outlived its relatively brief life span, however. The widespread social disruption resulting from the application of physical violence did more than complicate the establishment of a viable system of administration; it also left a legacy of anti-Western sentiment on which subsequent generations of nationalists were able to capitalize.

Belgian paternalism and the politics of decolonization. Belgian colonial rule bore traces of its Leopoldian pedigree: the irreducible tendency to treat Africans as childlike creatures and a firm commitment to political control and compulsion—on which Belgian paternalism was based—were both characteristic features of Leopoldian rule. The elimination of the more brutal aspects of the Congo Free State notwithstanding, Belgian rule remained conspicuously unreceptive to political reform. By placing the inculcation of Western moral principles above political education and welfare benefits above the apprenticeship of social responsibility, Belgian policies virtually ruled out all initiatives designed to foster political experience and responsibility.

Not until 1957, with the introduction of a major local government reform (the so-called *statut des villes*), were Africans given their first taste of democracy. By then the impact of social change had become apparent in the rise of a class of Westernized Africans (*évolués*) anxious to exercise their political rights beyond the urban arenas; the heavy demands made upon the rural masses during the years of the two world wars, coupled with the profound psychological impact of the postwar constitutional reforms introduced in neighbouring French-speaking territories, created a climate of social unrest suited for the development of nationalist sentiment and activity. The precipitating factor behind the political awakening of the Congolese masses was the publication in 1956 of a political manifesto calling for immediate independence. Penned by a group of Bakongo *évolués* affiliated to the Alliance des Bakongo (ABAKO), an association based in Léopoldville (now Kinshasa), the manifesto was the response of ABAKO to the ideas set forth by a young Belgian professor of colonial legislation, A.A.J. van Bilsen, in his "Thirty-Year Plan for the Political Emancipation of Belgian Africa." Far more impatient in tone, the ABAKO manifesto stated: "Rather than postponing emancipation for another thirty years, we should be granted self-government to-day."

The ABAKO manifesto

Under the leadership of Joseph Kasavubu, ABAKO transformed itself into a major vehicle of anticolonial protest; the ferment of nationalist sentiment quickly spread through the lower Congo region, and, in time, the nationalist contagion reached the rest of the colony. Scores of self-styled nationalist movements mushroomed almost overnight in each province. In the welter of political parties brought into existence by the *statut des villes,* the Congolese National Movement (Mouvement National Congolais; MNC) stood out as the most powerful vector of territorial nationalism. Although the MNC never disavowed its commitment to national unity (unlike ABAKO, whose appeal was limited to Bakongo elements), not until the arrival of Patrice Lumumba in Léopoldville, in 1958, did the party enter its militant phase.

The turning point in the process of decolonization came on Jan. 4, 1959, when anti-European rioting erupted in Léopoldville, resulting in the death of scores of Africans at the hands of the security forces. On January 13, the Belgian government formally recognized independence as the ultimate goal of its policies—a goal to be reached "without fatal procrastination, yet without fatal haste." By then, however, nationalist agitation had reached a threshold of intensity that made it virtually impossible for the Belgian administration to effectively control the course of events. To this growing turbulence the Belgian government responded by convening a Round Table Conference in Brussels, in January 1960, involving the participation of a broad spectrum of nationalist organizations. The aim was to work out the conditions of a viable transfer of power; the result, however, was an experiment in instant decolonization. Six months later, on June 30, the Congo formally acceded to independence, hurtling toward a self-induced apocalypse.

The Congo crisis. The triggering element behind the "Congo crisis" was the mutiny of the army (the so-called Force Publique) near Léopoldville on July 5, immediately followed by the intervention of Belgian paratroopers, ostensibly to protect the lives of Belgian citizens. Adding

Belgian intervention

to the confusion created by the collapse of the Force Publique, the constitutional impasse arising from the opposition between the president and the prime minister brought the machinery of government to a halt. President Kasavubu revoked Prime Minister Lumumba from his functions; Lumumba responded by dismissing Kasavubu. Meanwhile, on July 11, the country's richest province, Katanga, declared itself independent under the leadership of Moise Tshombe. The support given by Belgium to the Katanga secession gave a measure of credibility to Lumumba's claims that Brussels was trying to reimpose its authority on its former colony, and on July 12 he and Kasavubu appealed to UN Secretary-General Dag Hammarskjold for UN security assistance.

While intended to pave the way for the restoration of peace and order, the arrival of the UN peacekeeping force added yet another source of tension between President Kasavubu and Prime Minister Lumumba. The latter's insistence that the United Nations should use force if necessary to bring Katanga back into the fold of the central government met with categorical opposition from Kasavubu. Lumumba then turned to the Soviet Union for logistic assistance to send troops to the Katanga, at which point the Congo crisis became inextricably bound up with East-West issues.

As the process of fragmentation set in motion by the Katanga secession reached its peak, resulting in the breakup of the country into four separate fragments (Katanga, Kasai, Orientale Province, and Léopoldville), Chief of Staff Joseph Mobutu (later Mobutu Sese Seko) announced on Sept. 14, 1960, that the army would henceforth rule with the help of a caretaker government. The threat posed to the new regime by the Lumumbist forces was substantially lessened by the capture of Lumumba in December 1960, after his dramatic escape from Léopoldville, and his subsequent execution at the hands of the Tshombe government. Although Kasavubu's surrender of Lumumba to the Katanga secessionists was intended to pave the way for a reintegration of the province into the fold of the central government, not until January 1963, and only after a violent showdown between the European-trained Katanga gendarmerie and the UN forces, was the secession decisively crushed. It would take another year for the last bastion of secessionism, the pro-Lumumba Stanleyville government, to be brought to heel. Meanwhile, following the convening of parliament in Léopoldville, a new civilian government headed by Cyrille Adoula came to power on Aug. 2, 1961.

Even more than his inability to deal effectively with the Katanga secession, Adoula's decision to dissolve parliament in September 1963 brought his popularity to its lowest ebb. His move contributed directly to the outbreak of rural insurgencies, which, from January to August 1964, engulfed 5 provinces out of 21 and suddenly raised the ominous prospect of a total collapse of the central government. Because of its poor leadership and fragmented bases of support, the rebellion failed to translate its early military successes into an effective power apparatus; even more important in turning the tide against the insurgents was the decisive contribution made by European mercenaries in helping the central government regain control over rebel-held areas. For this, much of the credit goes to Moise Tshombe who, by July 10, had replaced Adoula as prime minister. Ironically, a year and a half after his defeat at the hands of the UN forces, the most vocal advocate of secessionism had suddenly emerged as the providential leader of a besieged central government.

1966 Katanga mutiny

Mobutu's regime. Mobutu's second coup, on Nov. 24, 1965, occurred in circumstances strikingly similar to those that led to the first—a struggle for power between the incumbent president, Joseph Kasavubu, and his prime minister, Moise Tshombe. Unlike Lumumba, however, Tshombe managed to leave the country unharmed—and determined to regain power. Rumours that the ousted prime minister was plotting a comeback from his Spanish retreat hardened into certainty when in July 1966 some 2,000 of Tshombe's former Katanga gendarmes, led by mercenaries, mutinied in Kisangani (formerly Stanleyville). Exactly a year after the crushing of the first mutiny another broke out, again in Kisangani, apparently triggered by the news that Tshombe's airplane had been hijacked over the Mediterranean and forced to land in Algiers, where he was held prisoner. Led by a Belgian settler named Jean Schramme, and involving approximately 100 former Katanga gendarmes and about 1,000 Katangese, the mutineers held their ground against the 32,000-man Congolese National Army (Armee Nationale Congolaise; ANC) until November 1967, when Schramme and his mercenaries crossed the border into Rwanda and surrendered to the local authorities.

The fragility of Mobutu's power base was again demonstrated in 1977 and 1978, when the country's main opposition movement, the Congolese National Liberation Front (Front de la Libération Nationale Congolaise; FLNC), operating from Angola, instigated two major invasions into Shaba (formerly Katanga). On both occasions external intervention from friendly governments—mostly from Morocco in 1977 and from France in 1978—saved the day, but at the price of untold casualties among Africans and Europeans. After the capture of Kolwezi in May 1978 an estimated 100 Europeans lost their lives, partly at the hands of the rebels and partly at the hands of the ANC. At any rate, and quite aside from the part played by the FLNC in spearheading the invasions, the sharp deterioration of the Zairian economy after 1975, coupled with the rapid growth of anti-Mobutu sentiment among the poor and the unemployed, were crucial elements in the background of the Shaba invasions.

The timing of the first Shaba invasion, 11 years after the creation of the Popular Movement of the Revolution (Mouvement Populaire de la Révolution; MPR), made plain the shortcomings of the single-party state as an agent of national integration and of Mobutism as a legitimizing formula. Officially described as "the nation politically organized," the MPR may be better seen as a weakly articulated patronage system. Mobutu's effort to extol the virtues of Zairian "authenticity" did little to lend respectability either to the concept or to the brand of leadership for which it stood. As befit his chiefly image, Mobutu's rule was based on bonds of personal loyalty between himself and his entourage. His hegemony was absolute, however, and extended to every level of the government. Whether Mobutu's decision, in April 1990, to lift the ban on opposition parties will decisively alter Zairian politics remains an open question. Although provision has been made for the introduction of a three-party system to replace the single-party rule of the MPR, the sheer brutality of the repression that followed in the wake of student protests at the University of Lubumbashi—resulting in the deaths of anywhere from 50 to 150 students, according to Amnesty International—leaves one to wonder whether Mobutu's professed commitment to democracy can be taken seriously. Thirty years after Zaire's independence and 82 years after the demise of the Congo Free State, the Leopoldian quality of the Mobutist state remained evident in its thoroughly centralized power apparatus, highly personalized style of governance, and readiness to use force whenever circumstances required. To the extent that history repeats itself, nowhere has it been more tragically repetitive than in the complex encounter of Zairian societies with the forces of modernity. (Re.L.)

For later developments in the history of Zaire, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

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Central America

Central America, in the strictest sense, is a region of North America that lies between Mexico and Panama, the latter a part of Colombia until 1903 and thus sometimes technically still included in South America. (Geologists and physical geographers often extend the boundary of Central America as far north as the Isthmus of Tehuantepec in Mexico.) Within the region are the countries of Guatemala, Belize (until 1973 British Honduras), Honduras, El Salvador, Nicaragua, and Costa Rica. Through custom Panama has come to be included as a part of Central America, however, and that convention is followed in this treatment.

Central America makes up most of the tapering isthmus that separates the Pacific Ocean from the Caribbean Sea. At its narrowest point, in Darién (Panama), the isthmus is only about 30 miles (50 kilometres) wide, and there is no location in Central America that is more distant than 125 miles (200 kilometres) from the sea. An arc joining Central America's most distant points would be about 1,140 miles long. The seven countries of the region have a land area of 202,265 square miles (523,865 square kilometres), and together they are only about half the size of neighbouring Colombia and about one-fourth that of adjoining Mexico.

Anthropologists often use the word Mesoamerica to define those areas that were the seats of high civilizations in pre-Columbian time; a line drawn from the Gulf of Honduras in the Caribbean Sea south to the Gulf of Fonseca in the Pacific Ocean and as far as the Gulf of Nicoya (Costa Rica) would separate Mesoamerica to the north and west from the rest of Central America. Cultural geographers include Central America as part of a larger region, Middle America, which includes Mexico and the Caribbean islands. Middle America is commonly divided into the Mainland, which accommodates most of Central America, and a circum-Caribbean region, called the Rimland, which includes the coastal lowlands of eastern Central America. In the Rimland the indigenous peoples have been largely swamped by immigrant populations, notably from northwestern Europe, Africa, and more recently the United States. Mainland Central America, on the other hand, has a stronger American Indian and part American Indian presence, as well as ways of life that still owe much to pre-Columbian and Spanish colonial times.

Major physical divisions

Central America is a fragmented region, a fine-grained mosaic of different landscapes; many parts are still physically isolated from each other. Highlands predominate over lowlands, steep slopes over flatlands. Altitude mitigates the climatic effects of a tropical latitude. Although the sea is close, few have chosen to live by it; the bulk of the population lives between the 3,000-foot (900-metre) and 8,000-foot (2,400 metre) lines of altitude and enjoys the lower temperatures and the freedom this provides from the diseases of the coastal lowlands. History and geography have dictated that most Central Americans occupy the highlands close to the Pacific and have eschewed the Caribbean lowlands. The majority of the eastern half of the isthmus remains under forest and empty of people. In contrast, in western Central America there are areas of intense population pressure on the land, and little of the natural vegetation has survived.

The brief political unity of Central America after independence from Spain lasted no longer than two decades before the provincialism of today's states asserted itself. Subsequently, the countries of Central America developed stronger links with the foreign nations providing markets for their export commodities, particularly the United States, than with their neighbours. Attempts to build a new political and economic community in the 1960s and '70s stalled in the financial crisis that followed. The aftermath of the crisis, the widespread civil unrest, and the political divisions of the 1980s increased the disunity and the internal contrasts within the isthmus. Central America remains a region of dichotomies, but perhaps the most glaring is the disparity between the poverty of the many and the wealth of a few.

This article is divided into the following sections:

THE REGION

Physical and human geography

THE LAND

Regional tectonics

Relief, drainage, and soils. Widespread volcanic activity and the frequency of severe earthquakes testify to the tectonic instability of many parts of Central America. This instability is produced by the interactions between the four relatively small crustal plates that provide the tectonic setting of the region. The northeastern edge of the Cocos Plate runs parallel to the Pacific coast some 60 to 125 miles offshore. It abuts on the southwestern edge of the Caribbean Plate, whose northern edge is marked by three fault zones that extend across southern Guatemala to include the active Cayman trough in the Caribbean Sea. North of the Cocos Plate to the west and north of the Caribbean Plate to the east is the North American

Plate, which is relatively stable and has allowed the steady accumulation of the limestones of the Yucatán Peninsula. The eastern edge of the Cocos Plate is in contact with the Panama block of the larger Nazca Plate.

It is the interaction between the Caribbean and the Cocos plates that is the most important, since the Cocos Plate is moving northeastward relative to the Caribbean Plate at, by geologic standards, the lightning speed of 33 feet per century. The movement is accommodated by subduction of the Cocos Plate where it meets the Caribbean Plate along the line of the submarine Middle America Trench and by the breaking up of the Caribbean Plate into blocks ahead of the trench. Subduction of the Cocos Plate brings with it uplift of land beyond the front; more dramatically, this results in subterranean pressure being released, which reverberates as earthquakes and volcanic eruptions at the surface. Earthquakes in Central America have proved much more destructive of life and property than have volcanic outbursts, although the region is notorious in both respects.

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A small farm nestled in the rugged highlands of central Honduras.

Four-fifths of Central America is hilly or mountainous, and areas of flatland away from the coasts are restricted. One familiar image of Central America is provided by the line of more than 40 volcanoes that tower above the Pacific coast for 900 miles from Tacaná (13,428 feet) on the Guatemalan-Mexican border to Turrialba (10,955 feet) in Costa Rica. The summit of Tacaná's neighbour Tajumulco (13,845 feet) is the highest point in Central America. In general the volcanic chain forms the watershed between the Caribbean and the Pacific. Those rivers draining eastward are longer than the westward-flowing streams and are well served with rainwater, but few are navigable far upstream. Although a quarter of the volcanoes are extinct and half are dormant, the remainder together provide the most active volcanic belt in the Americas. Fuego (12,346 feet), some 20 miles from Guatemala City, is one of the largest and has erupted four times since 1880. Other active volcanoes include Santa María in Guatemala, Izalco in El Salvador, Las Pilas-Cerro Negro and Concepción in Nicaragua, and Poás and Irazú in Costa Rica. Lava flows have caused great damage by impounding rivers until they burst their banks, but volcanic dust has rejuvenated local soils downwind. As spectacular as the great cones are the cliff-rimmed lakes or calderas—such as Lake Atitlán in Guatemala and Lake Ilopango in El Salvador—formed by volcanic collapse.

The calderas

The narrow Pacific plain (some 30 miles wide) southwest of the volcanoes is made of overlapping fans of river alluvium interspersed in El Salvador and Nicaragua by volcanic hills; the coastal sandbars and lagoons are broken only by the downfaulted Gulf of Fonseca. To the northeast of the volcanic axis at the point where Honduras, El Salvador, and Guatemala meet is the geologic heartland of Central America, the Chortis block, an area of Paleozoic metamorphic and igneous rocks overlain by moderately deformed Mesozoic and Cenozoic marine and continental sediments and by volcanic rocks.

In the far north at the foot of the Yucatán Peninsula is the low, rolling karstic country of the Petén, broken by lake- and swamp-filled depressions (*bajos*) such as Lake Petén Itzá in Guatemala. South of the Petén the landscape has been fashioned into a series of parallel, arcuate ranges and depressions trending east-west in western Guatemala and northeast-southwest in southern Belize, eastern Honduras, and Nicaragua, mirroring geologic structures of Cretaceous origin. The Cockscomb Range (3,675 feet high) of the Maya Mountains in Belize and the hills of the southern Petén are minor northern outposts of these ranges.

There are three major mountain ranges in Guatemala: the high, flat plateau (llanos) of the Sierra los Cuchumatanes (9,800 feet) in the northwestern part of the country and in the east the broken lower Sierra de las Minas (10,200 feet) of Verapaz and the knife-edge ridges of the lower Sierra de Merendón. Guatemala's major river, the Motagua, occupies a structural depression between the last two ranges. Steep-sided river valleys and deep canyons make overland access difficult.

The parallel ranges of Honduras and northern Nicaragua are broken by a number of small, alluvium-filled, fault-guided basins that run obliquely to the Caribbean coast, as do the rivers. In northern Honduras the crystalline bedrock impregnated with mineralized veins has weathered, producing a rugged landscape (with pockets of alluvial gold). In northeastern Honduras and Nicaragua east of Lakes Managua and Nicaragua the older rocks are buried under a thick covering of Tertiary volcanic material cut into by the numerous rivers that flow eastward across almost the full width of the isthmus. They have created a series of hilly basins separated by high ridges (3,300 to 6,600 feet) whose amplitude of relief declines eastward toward the Mosquito coastal plain. Here the rivers flow in shallow trenches, and their basins include a series of modest escarpments. This section of the Caribbean Plain is the most extensive of Central America's lowlands and is underlain by Pliocene marine gravels and sandy clays that unfortunately produce infertile lateritic soils. The Mosquito Coast is one of deltas, sandbars, and lagoons.

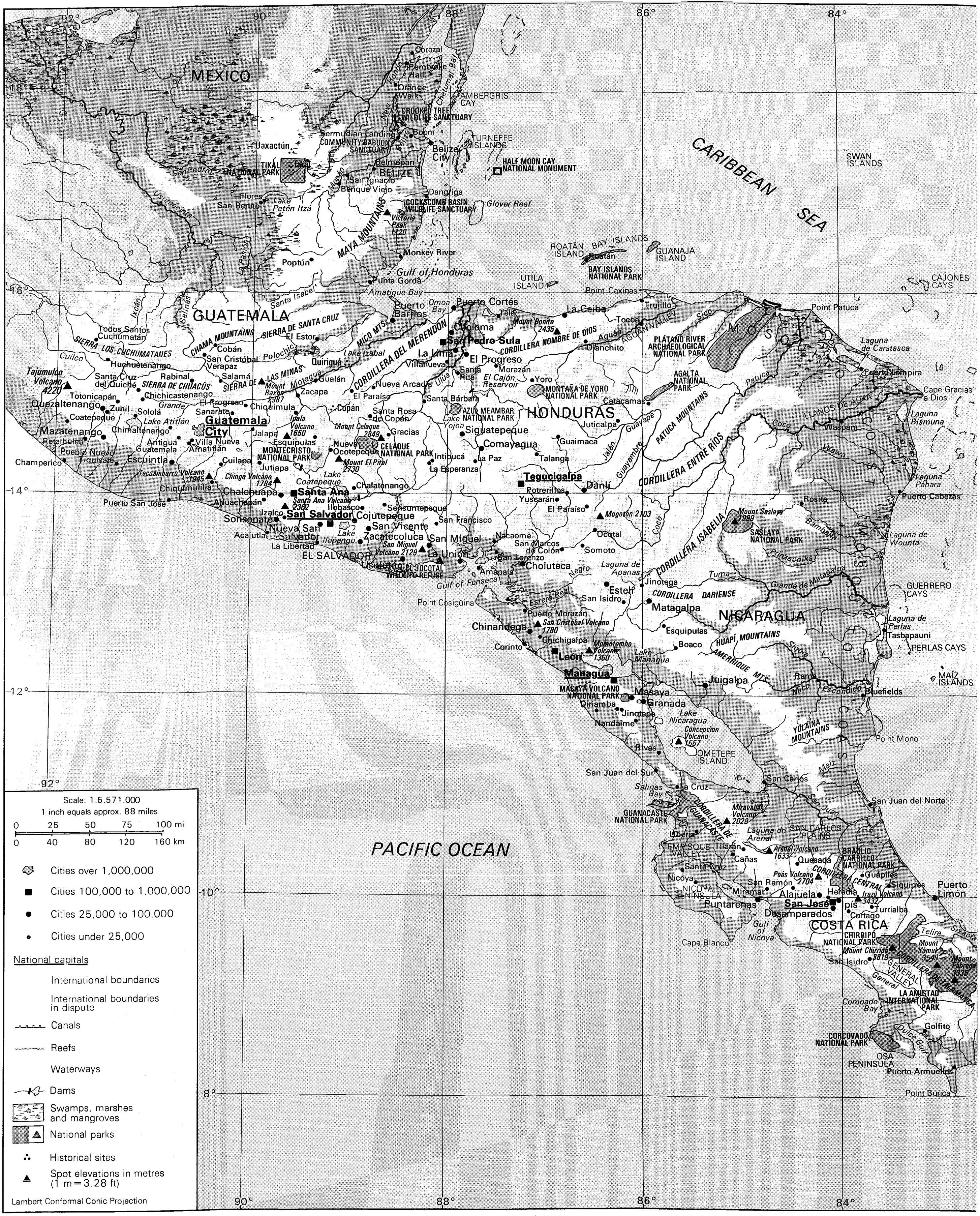
The Mosquito Coast

Southward the highlands end abruptly where a major rift valley in Nicaragua separates them from the volcanoes farther to the southwest. This rift valley runs diagonally across the Central American isthmus from its submarine expression as the Gulf of Fonseca, widening inland, where it is partly flooded by the fresh water of Lakes Managua and Nicaragua, before continuing as the valley of the San Juan River.

The San Juan is an exceptional river for Central America, having proved itself useful as a navigable routeway. It is about 125 miles from the point where the river drains Lake Nicaragua to the Caribbean Sea. The continental divide to the west of the lake is only 165 feet above the Pacific and only about 15 miles from the shore. These short distances and the low divide make the river-lake link an attractive route between the two oceans, and it was much used in the 19th century. It was indeed the route favoured by the first (1876) of the U.S. commissions investigating alternative transisthmian canal sites, a choice that ultimately fell to Panama.

South of the rift, in Costa Rica, is a portion of Cretaceous oceanic crust overlain by later marine and volcanic rocks that suffered severe deformation during late Tertiary and early Quaternary times and subsequent volcanic activity. Overlooking the rift are the high, volcanic Cordillera Central and northern Cordillera de Talamanca, intercalated by highland basins. The major part of the Cordillera de Talamanca is a granitic intrusion that blocks off the southern end of the volcanic axis. This is the most massive of the Central American mountain ranges, with 10 peaks rising above 9,800 feet, which were sculpted by glaciers in Pleistocene times. Slopes are steep, road building expensive, and landslides easily provoked; it is a refuge for small Indian groups even to this day. The Pacific face of the

Cordillera de Talamanca


MEXICO
GUATEMALA
BELIZE
HONDURAS
EL SALVADOR
NICARAGUA
COSTA RICA
CARIBBEAN SEA
PACIFIC OCEAN
92°
90°
88°
86°
84°
18°
16°
14°
12°
10°
8°
Corozal
Pembroke Hall
Orange Walk
Chetumal Bay
New River
AMBERGRIS CAY
CROOKED TREE WILDLIFE SANCTUARY
Bermudian Landing
COMMUNITY BABOON SANCTUARY
Boom
Belize City
TURNEFFE ISLANDS
HALF MOON CAY NATIONAL MONUMENT
Uaxactún
TIKAL NATIONAL PARK
Tikal
San Pedro
Belmopan
San Ignacio
Benque Viejo
Flores
San Benito
Lake Petén Itzá
Usumacinta
Mopan
Dangriga
COCKSCOMB BASIN WILDLIFE SANCTUARY
Glover Reef
Victoria Peak 1120
MAYA MOUNTAINS
La Pasión
Poptún
Monkey River
Gulf of Honduras
Punta Gorda
Amatique Bay
SWAN ISLANDS
ROATÁN ISLAND
BAY ISLANDS
Roatán
GUANAJA ISLAND
BAY ISLANDS NATIONAL PARK
UTILA ISLAND
Point Caxinas
CAJONES CAYS
Santa Isabel
Ixcán
Salinas
Todos Santos Cuchumatán
SIERRA LOS CUCHUMATANES
CHAMA MOUNTAINS
SIERRA DE SANTA CRUZ
El Estor
Cobán
San Cristóbal Verapaz
Polochic
Lake Izabal
Cuilco
Huehuetenango
Tajumulco Volcano 4220
Santa Cruz del Quiché
Rabinal
Salamá
SIERRA DE CHUACÚS
SIERRA DE LAS MINAS
Mount Raxón 2987
Motagua
Quiriguá
Gualán
Zacapa
Totonicapán
Chichicastenango
Grande
El Progreso
Quezaltenango
Zunil
Sololá
Sanarate
Chiquimula
Coatepeque
Lake Atitlán
Guatemala City
Mazatenango
Chimaltenango
Jalapa
Ipala Volcano 1650
Retalhuleu
Antigua Guatemala
Villa Nueva
Amatitlán
Esquipulas
Pueblo Nuevo Tiquisate
Champerico
Escuintla
Cuilapa
Jutiapa
MONTECRISTO NATIONAL PARK
Tecuamburro Volcano 1945
Chingo Volcano 1784
Chiquimulilla
Puerto San José
Puerto Barrios
Omoa Bay
Puerto Cortés
Tela
La Ceiba
Trujillo
Tocoa
MICO MTS.
CORDILLERA DEL MERENDÓN
Choloma
San Pedro Sula
Mount Bonito 2435
CORDILLERA NOMBRE DE DIOS
Aguán
AGUÁN VALLEY
Sico
Olanchito
La Lima
Villanueva
El Progreso
Santa Rita
Ulúa
Morazán
El Cajón Reservoir
Yoro
MONTAÑA DE YORO NATIONAL PARK
Nueva Arcadia
El Paraíso
Copán
Santa Rosa de Copán
Santa Bárbara
Lake Yojoa
AZUL MEAMBAR NATIONAL PARK
Mount Celaque 2849
Gracias
CELAQUE NATIONAL PARK
Nueva Ocotepeque
Mount El Pital 2730
Intibucá
La Esperanza
La Paz
Siguatepeque
Comayagua
Guaimaca
Talanga
Catacamas
Juticalpa
Guayape
Jalán
Guayambre
PATUCA MOUNTAINS
CORDILLERA ENTRE RÍOS
Patuca
PLÁTANO RIVER ARCHAEOLOGICAL NATIONAL PARK
AGALTA NATIONAL PARK
MOSQUITO COAST
Point Patuca
Laguna de Caratasca
Puerto Lempira
Cape Gracias a Dios
LLANOS DE AUKA
Coco
Waspam
Laguna Bismuna
Wawa
Laguna Pahara
Puerto Cabezas
Tegucigalpa
Potrerillos
Danlí
Yuscarán
El Paraíso
Mogotón 2103
Lake Coatepeque
Chalatenango
Chalchuapa
Santa Ana
Ahuachapán
Santa Ana Volcano 2362
Ilobasco
Sensuntepeque
Izalco
San Salvador
Cojutepeque
Sonsonate
Nueva San Salvador
San Vicente
San Francisco
Acajutla
La Libertad
Lake Ilopango
Zacatecoluca
San Miguel Volcano 2129
San Miguel
Usulután
La Unión
EL JOCOTAL WILDLIFE REFUGE
Gulf of Fonseca
Amapala
Nacaome
San Marcos de Colón
San Lorenzo
Choluteca
Ocotal
Somoto
Laguna de Apanas
Negro
Jinotega
Mount Saslaya 1999
SASLAYA NATIONAL PARK
CORDILLERA ISABELIA
Bambana
Laguna de Wounta
Prinzapolka
Rosita
Tuma
Grande de Matagalpa
Estelí
San Isidro
CORDILLERA DARIENSE
Matagalpa
Estero Real
Point Cosigüina
Puerto Morazán
Chinandega
San Cristóbal Volcano 1780
Chichigalpa
Corinto
León
Momotombo Volcano 1360
Esquipulas
Boaco
HUAPÍ MOUNTAINS
AMERIQUE MTS.
Siquia
Lake Managua
Managua
Juigalpa
Rama
Mico
Escondido
Bluefields
Laguna de Perlas
Tasbapauni
PERLAS CAYS
MAÍZ ISLANDS
GUERRERO CAYS
MASAYA VOLCANO NATIONAL PARK
Masaya
Granada
Diriamba
Jinotepe
Nandaime
Lake Nicaragua
Concepción Volcano 1557
OMETEPE ISLAND
Rivas
YOLAINA MOUNTAINS
Point Mono
Maíz
San Juan del Sur
San Carlos
San Juan
San Juan del Norte
Salinas Bay
La Cruz
GUANACASTE NATIONAL PARK
CORDILLERA DE GUANACASTE
Miravalles Volcano 2028
Liberia
Laguna de Arenal
SAN CARLOS PLAINS
TEMPISQUE VALLEY
Tilarán
Arenal Volcano 1633
Cañas
Quesada
BRAULIO CARRILLO NATIONAL PARK
Santa Cruz
Nicoya
NICOYA PENINSULA
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CORDILLERA CENTRAL
San Ramón
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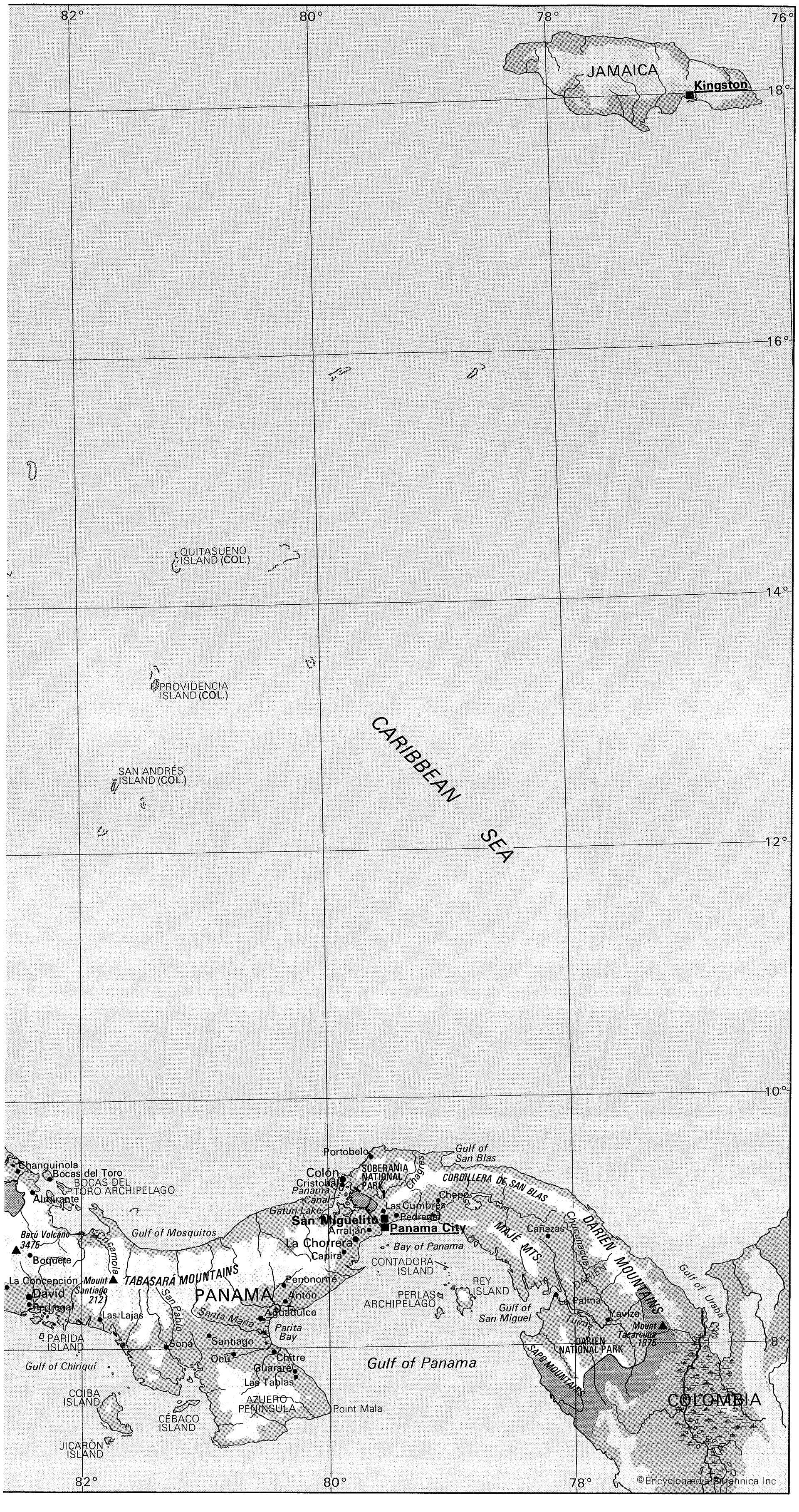
GENERAL VALLEY
General
LA AMISTAD INTERNATIONAL PARK
Coronado Bay
Golfito
Dulce Gulf
CORCOVADO NATIONAL PARK
OSA PENINSULA
Puerto Armuelles
Point Burica
Scale: 1:5,571,000
1 inch equals approx. 88 miles
0 25 50 75 100 mi
0 40 80 120 160 km
Cities over 1,000,000
Cities 100,000 to 1,000,000
Cities 25,000 to 100,000
Cities under 25,000
National capitals
International boundaries
International boundaries in dispute
Canals
Reefs
Waterways
Dams
Swamps, marshes and mangroves
National parks
Historical sites
Spot elevations in metres (1 m = 3.28 ft)
Lambert Conformal Conic Projection



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Cordillera de Talamanca is a fault scarp that overlooks the structural basin of the General River and the low coast ranges beyond.

Volcanism resumes in western Panama, where Barú (Chiriquí) Volcano (now inactive) reaches 11,401 feet, and highly dissected volcanic ranges declining in height form the spine of the country to beyond the Panama Canal. The mountainous Azuero Peninsula extends southward into the Pacific and separates the Gulf of Panama from that of Urabá; rolling lowlands made fertile by ejecta from the nearby volcanoes lie between the Azuero Peninsula and the central cordillera of Panama to the north.

Climate. The climates of Central America are essentially tropical, tempered by proximity to the sea, by altitude, by latitude, and by local topography; in consequence, they may vary substantially over short distances. There is a net east-to-west flow of air across Central America, with pressure higher on the Caribbean side and lower on the Pacific side. The adjoining sea surfaces are warm, averaging about 81° F (27° C) throughout the year, and the Caribbean current in the east and the equatorial countercurrent in the Pacific carry their warmth northward beyond the isthmus. The atmosphere tends to be moisture-laden, but there is more rainfall to the south and the Caribbean side of the isthmus is wetter than the Pacific side.

Altitudinal climatic regions

The effect of altitude on temperature has long been popularly recognized in Central America. The *tierra caliente* ("hot land") is commonly defined as land below 3,300 feet in altitude, with mean annual temperatures exceeding 72° F (22° C); the *tierra templada* ("temperate land") between 3,300 and 6,500 feet, with mean annual temperatures ranging between 60° and 72° F (16° and 22° C); the *tierra fría* ("cold land") between 6,500 and 9,800 feet, with mean annual temperatures below 46° F (8° C); and the *tierra helada* ("frozen land") above that. The relationship between altitude and temperature is not a linear one: the higher the altitude the more rapid the falloff in temperature as the absolute humidity declines. Probably the more important climatic contribution of altitude in Central America is made when topographic barriers athwart the airflow force humid air to rise, perhaps cooling it below its dew point, shrouding the land in cloud, and dousing the higher slopes with rain; as the air descends the leeward side, it brings warmer and drier weather.

The seasonal movement of the sun means that the intertropical convergence zone (the ITCZ, marked by the axis of the equatorial low-pressure belt of the doldrums) is at its most northerly extent in late summer, when it lies to the north of Panama along approximately latitude 10° N in the east, veering northwestward in the vicinity of San José to bisect Lake Nicaragua and run along the crestline of the Pacific coastal ranges. It brings with it unstable weather marked by convectional rainfall, thunderstorms, and even hail in the Guatemalan highlands. To the south of the zone the surface winds are southwesterlies; to the north the trade winds blow from the east. The southern part of Central America is under the influence of the ITCZ for much of the year. Weak, ill-defined easterly waves in the flow of air on the north side of the ITCZ form over the Caribbean several times during a normal summer and move westward into Central America, bringing *temporales,* days with clouds and heavy rain.

The distinction between a drier winter and a wetter summer is more marked in the north. Furthermore, the seasonal reversal of winds on the Pacific side means that they are frequently onshore in summer, causing wetter weather, and offshore in winter, bringing drier conditions. Central America, north of latitude 12° N, is occasionally visited in winter by the outer edges of cold polar air masses that bring frontal storms, high winds (*nortes*), cool temperatures, and generally unsettled weather. Between May and November, and particularly from August to October, the Caribbean coast of northern Central America is exposed to the possibility of hurricane damage. Hurricanes generally increase in frequency northwestward. The tropical storms that form along the ITCZ off the Pacific coast are less numerous but are at their most frequent in September.

Precipitation

The combination of maritime, orographic, seasonal, and latitudinal influences and diurnal fluctuations gives great local variety to the climates of Central America. Tropical wet climates characterize the Caribbean coastal plains and piedmont, the Petén, Darién, and the Pacific coast between and including the Osa and Azuero peninsulas. In these areas monthly temperatures exceed 64° F (18° C) throughout the year, rainfall exceeds about 80 inches (2,000 millimetres), and there is no seasonal drought. The wettest area is around San Juan del Norte (Greytown), Nicaragua, where the annual rainfall averages more than 236 inches (6,000 millimetres), with about 6 inches (150 millimetres) falling in the driest month (March) and 31 inches (800 millimetres) in the wettest (November). The sunnier months are the drier months; in April, for instance, Cristóbal at the Caribbean end of the Panama Canal records three-fourths of the total sunshine possible, while in July and November clouds reduce the proportion to 38 percent.

The lowlands fronting the Gulf of Panama, the Pacific half of Nicaragua including the lake basins, the Gulf of Fonseca, El Salvador, and the Pacific slopes of Guatemala are all within the *tierra caliente* but experience a distinct dry season of four to six months. Only about 15 percent of the rain on the Pacific coast falls between November and April, less than in each of the three wettest months (September, October, and June) individually. Annual rainfall on the Pacific coast ranges between 39 and 79 inches (1,000 and 2,000 millimetres). In highland Central America mean temperatures are lower; rainfall tends to be more intense than in the corresponding lowlands, and the rain shadow effect can be strong.

Plant life. The vegetation of Central America reflects its intercontinental situation, the major influence of climate, the more localized impact of soil quality, topography, and the selective and accelerating modifications made by humans. Together these factors produce a great variety of landscapes. The tropical landmass of South America and, to a lesser extent, adjoining North America have added a rich immigrant flora to that indigenous to the isthmus.

A generous supply of solar energy, and the relative absence of environmental stress in lowland areas over recent geologic time has encouraged plant speciation. The lowland rain forests have strong similarities with the selva of South America, and the evergreen forests that clothe the lower Pacific slopes from Costa Rica northward remain essentially South American in floral composition; above 3,300–5,200 feet, affinities with North America are closer and the number of species fewer. For example, the important trees of the forested highlands of northern Central America are the pine (*Pinus*) and the oak (*Quercus*), as in Mexico. The pines extend no farther south than north-central Nicaragua; in the Cordillera de Talamanca in Costa Rica the oaks are partnered by conifers of South American provenance. Above the altitudinal tree line the bunchgrass landscape of Guatemala has Mexican, even U.S., associates; that of Costa Rica resembles the Andean *páramo* (barren plain).

The rain forest

Tropical rain forests are the natural vegetation of the eastern half of lowland Central America, possibly excluding northeastern Honduras, and of Pacific Panama. They are characterized by a broad variety of distinct tree species that form a closed crown forest of four or five stories; undergrowth is restricted to, for example, shade-tolerant palms, figs (*Ficus*), and large-leafed aroids (*Xanthosoma*). Vines (lianas) and epiphytes (for example, Spanish moss [*Tillandsia*]) are common. The trees are broad-leaved and evergreen; their trunks are often smooth, straight, and buttressed, and they rarely grow in pure stands. The forest canopy commonly reaches 130 feet above the ground and varies in composition from place to place. Along riverbanks and wherever the crown is broken and sunshine reaches the ground, a dense jungle may grow; otherwise the forest floor is relatively open. Among the rain forest trees that remain commercially attractive are the breadnut, mahogany, sapodilla, or chicozapote (whose latex, chicle, is the basis of chewing gum), tropical cedar in the Petén and in Belize, various dyewoods from the Mosquito Coast, cashew in the Dulce Gulf area of Costa Rica, and rubber tree in Panama.

The original vegetation of northeastern Honduras and northeastern Nicaragua and the Caribbean coastal plain of central and southern Belize is probably pine (*Pinus caribaea*) savanna. The rainfall would seemingly support rain forest, but porous quartzitic soils permit only a dry drought-tolerant evergreen formation. The pines may grow to 115 feet, sometimes in stands attractive to timber companies, but more usually widely spaced in a parklike landscape; elsewhere the trees may be small and interspersed with shrubs in a savanna emulsion cut by ribbons of riparian hardwoods, as is the case in the limited savanna that separates the Nicoya Peninsula from the rest of Guanacaste in Costa Rica. Human activity has extended these savannas. In many places along both coasts almost impenetrable mangrove swamps of salt-tolerant trees thrive, some limited to either the Caribbean or the Pacific sides of the isthmus.

The natural cover of the Pacific slopes from the Nicoya Peninsula northward is a forest that loses its leaves during the dry season (*monte alto*). The forest declines in height from about 100 feet, and the number of xerophytic species present and the proportion of shrubs to trees rise as the altitude and the intensity of the dry season increase.

Montane forest

The high interior of Guatemala and Honduras and of Nicaragua east of the lakes, as well as the central cordillera of Costa Rica and western Panama, are under montane forest, that is, forests where high altitude, notably through decreased temperature, imposes various structural limitations not existing at lower elevations. On the Caribbean side the tropical lowland rain forest blends upward into an evergreen montane forest, and the number of stories drops to two once the *tierra templada* is reached. With increased altitude the upper story becomes lower (about 65 feet), and in the *tierra fría* mosses and tree ferns become more important.

At about 6,500 feet (in southwestern Costa Rica 4,900 feet) cloud forest, dripping wet throughout most of the year, creates an unusual landscape in which evergreen oaks and laurels, epiphytes (especially bromeliads), and orchids are significant. Above the cloud forest oaks and euphorbias grow among an undergrowth of barberry, myrtle, and brightly coloured herbs, which in turn give way to a tangled evergreen woodland of shrubs and low trees. Above 9,800 feet in Costa Rica is the *páramo,* open areas of evergreen herbs, shrubs, and berries of the saxifrage, heath, and vaccinium families, myrtle, and sphagnum moss.

Animal life. Central American wildlife is similar to but not as rich as that of South America. A number of species also have North American ties, and some endemic species have evolved from western Panama, eastern Costa Rica, and the highlands of Guatemala, Honduras, and northern Nicaragua. Uncontrolled hunting, the demand of external markets, the destruction of natural habitats, and the pressures of a growing human population, however, have taken a severe toll on Central America's wildlife; many of the region's better-known inhabitants are endangered or seriously depleted species, and some are in imminent danger of extinction.

The jaguar, ocelot, jaguarundi, and margay, all South American cats, may be found in the isthmus but are becoming rarer. The puma, gray fox, and coyote have colonized from North America, the latter only as far as northern Central America. The raccoon has used the isthmus as a corridor into South America; and, in reverse, the armadillo and the opossum have made their way into the United States, and the quiet kinkajou and giant anteater have used the isthmus as a route into Mexico. There are many overlaps of northern (Nearctic) and southern (Neotropical) fauna in Central America. There are also some unaccountable breaks in the chain. For example, there are bears (the spectacled bear) in Mexico and Panama but none in between.

Forest fauna

The Central American forests are relatively sparsely populated with mammals, generously populated with reptiles, and extremely rich in birds and insects. The shaded forest floor supports few grazing ungulates, the peccary, the tapir, and the deer being among these; all are sought for food. Other forest mammals are arboreal. For example, monkeys are represented throughout Central America by the howler and the spider; the capuchin is found as far north as Honduras. The red-backed squirrel monkey of Panama and Costa Rica (where it is protected) has been heavily sought for the biomedical industry.

Water animals and reptiles are numerous. The amphibious river otter is on the endangered list. One of the more unusual mammals is the manatee, which inhabits lagoons and estuaries; it has been decimated through hunting, however. The landlocked shark found in Lake Nicaragua is one of the more interesting freshwater fish. Among the more familiar endangered reptiles are sea turtles. The tortoiseshell of the hawksbill and the meat of the green turtle are particularly prized, and the eggs of others are eaten. Lizards are well represented and include the gecko; the large common iguana and the related black iguana are sold live for meat in local markets. The caiman is found only as far north as latitude 14° N. Snakes are in abundance. The boas are represented by boa constrictors, but none of the giant anacondas of South America penetrate the isthmus. Three-fourths of the snakes belong to the Colubridae family. There is a rich fauna of amphibia, including tree frogs.

Central America's resident bird population is enriched by migratory flocks, their paths converging as the isthmus tapers southward. Parrot species are numerous. Some of the many genera of hummingbirds reach into the highest parts of the isthmus. The massive monkey-eating eagle is one of the largest birds of the tropical rain forests. The quetzal, whose feathers were so highly sought after by Maya, Aztec, and British Victorians alike, is now confined to the more remote parts of the cloud forests. The toucans penetrate as far north as Mexico.

Insects abound and are particularly obvious along rivers; colourful butterflies, armies of ants, and swarms of mosquitoes are witness to many other less visible insects.

Settlement patterns. Central America can be roughly divided into a very lightly peopled Caribbean half and a slightly more populous Pacific coastal slope, separated by a sequence of relatively densely settled highland regions. Rural population densities of under 13 persons per square mile (5 persons per square kilometre) are the rule on the Caribbean side of the isthmus, and many of the small towns there are closer in form and organization to those of the old British West Indies than to those farther west. On the Pacific slope settlement has often been fairly recent. Brash towns built to a gridiron plan and offering a range of commercial services are fringed by self-built, one-story family houses in a rural setting of scattered plantations, ranches, or farms that include associated accommodations for labourers.

Rural densities

In contrast, rural population densities in excess of 260 per square mile are widespread on the fertile volcanic soils of the older-settled highlands, notably on the southwestern highlands of Guatemala, in much of El Salvador, around the lakes in western Nicaragua, and in the highland basins of central Costa Rica. These highlands supported large populations before the Spanish conquest, and the availability of their labour, a comfortable climate, and healthier living conditions helped confirm this pattern in colonial and postcolonial times.

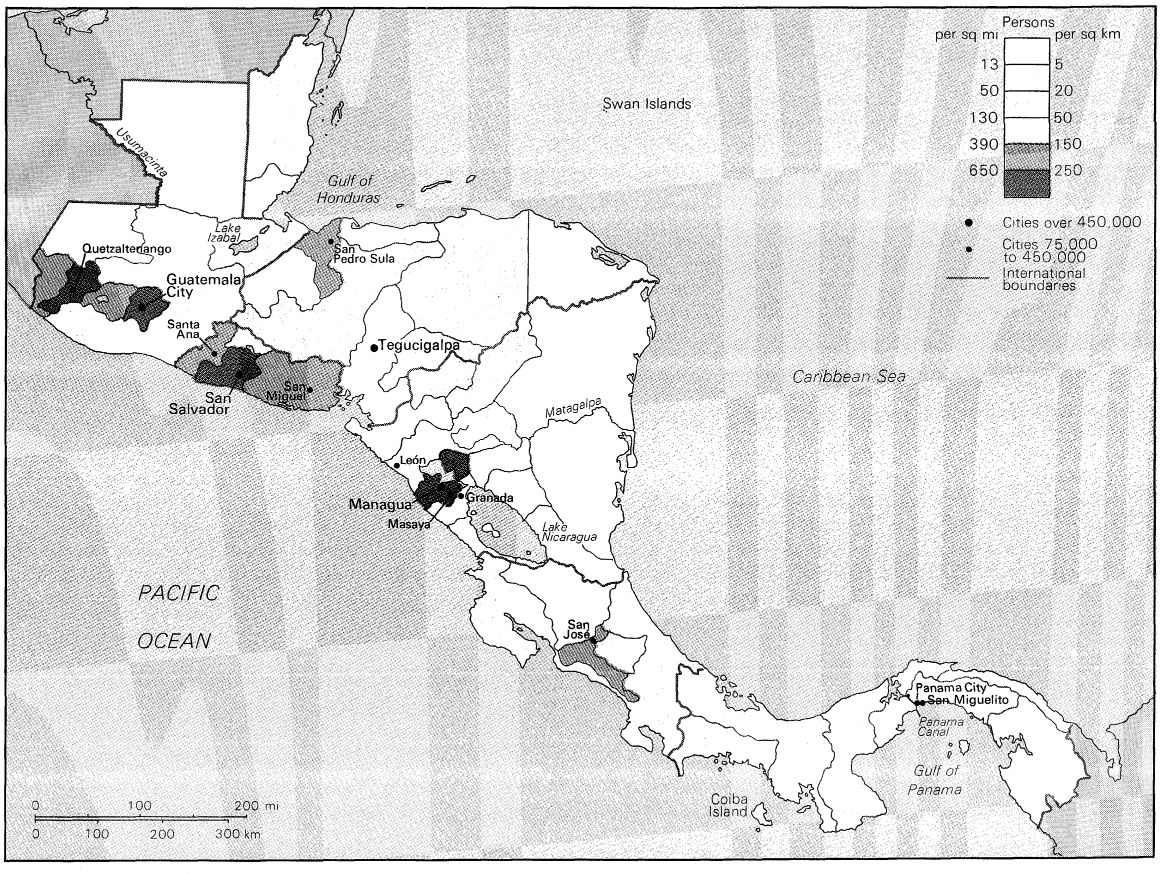
In the more heavily Indianized areas of Guatemala people live in individual homesteads scattered irregularly among their crops but with an association, and perhaps a second residence, in a meeting place or concourse, normally unoccupied but housing one or two civic or ceremonial buildings and whose main function is to serve as a periodic marketplace. Such concourses may be likened to the ceremonial centres of pre-Columbian times. Other highland peoples, of more varied ethnic origin, are more likely to live in small, compact towns whose structures are clearly planned and exhibit characteristics identifying their Hispanic origins. The town centre is likely to be an arcaded plaza bordered by an imposing church, administrative offices, public buildings, and the homes of some of the more important citizenry. A rectilinear grid of streets is aligned on the sides of the plaza. The farther from the centre, the lower the social class of the neighbourhood. This is the colonial mold that has been followed throughout highland Central America. More recently glass and steel office blocks and hotels have begun to change the skylines of the larger towns, and in some the original plaza has ceded its premier place to a more fashionable city centre in a newer part of town.

THE PEOPLE

Ethnic composition. At least two-thirds of Central Americans are of mixed race. Census figures estimate that 60 percent are of mixed European and American Indian blood (called ladinos in Guatemala and mestizos elsewhere), 5 percent are of mixed European and black origin (mulattoes), and 1 percent are of mixed American Indian and black descent (zambos). A further 20 percent is pure American Indian, or Amerindian, as they are often called in Middle and South America. Some 12 percent claims white European ancestry, while the descendants of Chinese and East Indian indentured labourers and others make up about 2 percent. Seven out of every eight persons classed as American Indians live in Guatemala, making up about 55 percent of that country's total population. One-third of the remaining American Indians live in adjoining Honduras, and another one-fourth in El Salvador; elsewhere, only in relatively empty Belize are American Indians a significant element (10 percent) in the population. Almost all the American Indians in Guatemala belong to branches of the Mayan family.

American Indians

Linguistic evidence suggests that the Maya core area was near Huehuetenango in northwestern Guatemala and that the first of a number of migrations began some four and a half millennia ago, resulting in the many distinct Maya tribal groupings present in the isthmus today. The Ameri-



Population density of Central America.

can Indians of northeastern Petén and northern Belize are members of the Yucatecan Maya (for example, the Itzá and Mopan), most of whom live on the Yucatán Peninsula in neighbouring Mexico. The northwest highlands of Guatemala are home to the Mamean people, including the Mam proper, the Ixil, and, straddling the Mexican border, the Kanjobalan, Chuj, and Jacaltec. The largest concentration of American Indian language speakers is that of the Quichéan peoples, who live in the midwestern highlands of Guatemala; they include the Quiché proper, the Cakchiquel, and the Tzutujil. The eastern highlands of Guatemala house the Pokomam, with a Quichéan language root, and include the Kekchí of Verapaz and southern Belize; the Chortí, who inhabit the area where Honduras, El Salvador, and Guatemala meet, share their linguistic roots with the American Indians of Tabasco in Mexico.

Costa Rica is the only country with a majority of residents descended from white Europeans. That country had only a small American Indian population in the 16th century and was unattractive to Spanish conquistadores, but it has been more attractive to later European migrants than have the other countries. Blacks are proportionally higher in numbers in Belize (40 percent) and Panama (14 percent), where West Indians were imported in large numbers to help build the Panama Canal.

Spanish linguistic dominance

Linguistic composition. Spanish is the dominant language of Central America and the official language in six of the republics; English is the lingua franca of much of the Caribbean coast and the official language of Belize.

Many of the American Indians are monolingual, some speak Spanish as a second language, and a considerable number speak more than one American Indian language. There has been an exchange of vocabularies among the languages. There are also various Cariban languages, with trans-Caribbean roots, spoken by the Black Caribs (*morenos*) and the small groups of Jicaque, Paya, Miskito, and Sumo of the Mosquito Coast and parts of Belize.

Religion. Roman Catholicism is the professed religion of six out of every seven Central Americans; almost all the remaining share various Protestant faiths. The conversion of the native peoples during colonial times was often accompanied by the absorption of Roman Catholicism into the rituals of pre-Hispanic beliefs and practices, and these remain today. More recently, evangelical Protestant sects have made substantial inroads into traditionally Roman Catholic communities, particularly in Guatemala and El Salvador, undermining the political importance of the Catholic hierarchy. There are small numbers of Muslims and Hindus among the East Indian population.

Demographic trends. The rapidly growing population of Central America reached 10 million in the early 1950s, 20 million in the late 1970s, and more than 28 million in mid-1988. About one-third of the people live in Guatemala, a sixth in El Salvador, another sixth in Honduras, an eighth in Nicaragua, a tenth in Costa Rica, and a twelfth in Panama; less than 1 percent of the people live in Belize.

Comparative vital statistics

The region retains one of the highest rates of population growth in the world, but it has been declining. The variables that determine demographics have been changing significantly; for example, crude death rates at 15 per thousand in 1965 were halved by 1987, while birth rates fell by 25 percent and net emigration increased. There are substantial national differences; for example, the relatively low death rates in Costa Rica, Belize, and Panama are half those in the other countries. Children born in Costa Rica and Panama can expect to live into their 70s, 10 years longer than those born in the northerly republics. Crude birth rates are also lower in Costa Rica and Panama than in El Salvador and the rest of the isthmus. More revealing are the contrasts in fertility rates; although falling, they are very high in Guatemala, Honduras, Nicaragua, Belize, and El Salvador but considerably lower in Costa Rica and Panama. The increased use of modern contraceptives is the most important factor influencing these variations.

Central Americans have become more mobile. In the past there was a net gain from migration—notably of West Indian labourers brought in to help build the Panama Canal and to work on the banana plantations of the largely empty Caribbean lowlands—but in more recent years there has been a net outflow of people from the isthmus. During the 1970s and '80s and into the '90s there was a substantial emigration of Central Americans—particularly to Mexico, the United States, and Canada—seeking escape from poverty, persecution, and the civil wars in their native lands.

Within Central America the greatest pressure to emigrate has been felt in El Salvador: 300,000 Salvadorans had migrated to neighbouring Honduras before the Soccer War of 1969, which resulted in the repatriation of some 130,000. More recently, the civil unrest of the 1980s displaced up to 1,000,000 people, including an estimated 500,000 Salvadorans who entered the United States, about 40 percent illegally. Tens of thousands of others migrated to Mexico, Guatemala, Honduras, and other countries in the region. There has long been a seasonal migration from Guatemala to pick the coffee beans of Mexico's Soconusco coast, and many have made the migration permanent.

Internal migration has made the population of Central America more urbanized. In 1960 only about one-third of the total lived in towns; by the late 1980s the proportion was more than 45 percent. Capital cities have attracted the greatest urban migration. About one-half of urban Costa Ricans live in metropolitan San José, for example; one-third of urban Panamanians live in Panama City; one-third of the townspeople of Nicaragua live in Managua; and a similar proportion of Guatemala's urban population lives in Guatemala City, the most populous city of the isthmus. But most of the towns are small, and urban life is strongly rooted in the local countryside.

Where there is a large American Indian population, as in northern and western Guatemala, rural traditions are particularly strong: two out of every three Guatemalans are officially classed as rural. Although the rural population of the isthmus has increased less rapidly than the urban and much of Central America's countryside remains uncrowded, there are a number of areas where the local pressure on land resources has become difficult if not impossible to contain. These areas include the uplands basins of western Guatemala, the core area of El Salvador, the Meseta Central of Costa Rica, and, to a lesser extent, the lakes region of Nicaragua. These are the originating areas of many of the internal migrants.

THE ECONOMY

Gross national product

The gross national product of Central America is small compared to more developed regions of the world. In the late 1980s it was, for example, on a par with that of Peru but only a fifth of that of Mexico. This suggests the small size of domestic markets in Central America and explains why the main economic focus over the years has been on producing commodities for export (or, in the case of Panama, providing international services). The performance of the Central American economies has closely mirrored that of the world, thanks to their export orientation. Thus, the overall trend of a 10-fold growth in real gross domestic product (GDP) between 1920 and 1990 has had imposed upon it the local impacts of the downturns of the Great Depression, World War II, and the 1980s, as well as the booms that followed both world wars and that of the 1970s.

The first export staple was coffee, firmly established on a commercial basis in Costa Rica by 1830 and in El Salvador, Nicaragua, and Guatemala by 1860; bananas joined coffee in 1900, and for the next half-century the two dominated trade and investment. Governments acquiesced with the interests of both the highland coffee growers, mainly nationals well integrated into local politics, and the U.S.-owned banana companies in the coastal lowlands. Accordingly, the governments wielded relatively little power; in contrast, by 1929 two fruit companies (the United Fruit Company and the Standard Fruit and Steamship Company) controlled all banana exports and most banana production in the isthmus. The addition of cotton, sugar, and beef to the mixture of export staples after World War II helped reduce the exposure of national economies to adverse changes in the prices of coffee or bananas but did not lessen the overall dependence upon external commodity markets.

The formation of the Central American Common Market (CACM) in 1960 gave a further impetus to economic diversification. By introducing import substitution policies and the manipulation of tariffs, national industries were encouraged to manufacture for the wider isthmian market. The enlarged market generated new employment, mainly in the towns, and helped encourage urban migration, and it created a new, if not very significant, regional trading pattern. In the 1970s government policies encouraged further diversification by promoting new exports. In addition, Panama's location at an international crossroads proved most useful in attracting offshore manufacturing and banking facilities. The effective demise of the CACM in 1980 (Panama and Belize, for different reasons, had never been members) reversed the limited progress made toward regional economic cooperation and helped highlight some of the national differences within the isthmus.

Causes of economic disruption

National circumstances have provided further departures from regional compatibility. The differing degrees of dependency upon particular commodities, for example, have been an important discriminant; another has been the greater encouragement given to import substitution policies in foodstuffs during the 1930s and in manufactured goods in the 1960s by Costa Rica, El Salvador, and Guatemala. Also important in several countries has been the variety of oppressive military governments and of open wars, guerrilla activity, and counterinsurgency measures.

These differences were compounded in the 1980s by changes in the terms of trade and the burden of huge external debts. The disbursed external debt of the isthmus rose fivefold in the 1970s and fivefold again in the '80s. The differing degrees of indebtedness and of ability to attract overseas assistance and investment helped widen the distinctions between the individual countries. By the end of the decade Nicaragua owed most, suffered the highest ratio of interest payments to export earnings, and was the least creditworthy. Panama and Costa Rica were more indebted, but payments due to service public debt were more onerous for Honduras and Guatemala. Overseas aid, which included military assistance, was made more readily available to Costa Rica, El Salvador, and Honduras but was essential to all.

Per capita production in Central America grew between 1960 and 1980 but actually fell during the 1980s. It was the poor who suffered most from this reversal. At the beginning of the decade there were already great disparities between the incomes of the poor and the rest of society in Central American countries. Thus, the poorest 20 percent of the isthmian population had only 3 percent of the national product at their disposal, while the richest 20 percent had 56 percent of the national income. The economic crisis of the 1980s exaggerated these differences.

Revised credit policy

The important roles foreign institutions and overseas markets have played in steering Central American economies in the past has meant that the growing influence assumed by the government of the United States and the various international financial agencies (notably the International Monetary Fund and the World Bank) over the direction of these nations' policies in the aftermath of the 1980s crisis was more readily accommodated than perhaps elsewhere in Latin America. The generally open, export-led nature of Central American economies has made them better able to adjust to the policies imposed by their modern creditors. These creditors, for example, have insisted on more effective and equable fiscal policies and a more efficient use of agricultural land, matters that were of relatively little concern to governments in earlier days.

Agriculture. Agriculture looms large in the lives of Central Americans. More than one-half of the people continue to live in the countryside, and many of those who live in towns continue to earn their livelihoods directly or indirectly from the land. The contribution agriculture makes to Central America's GDP (about 25 percent) is considerably higher than that of Latin America as a whole. About half the agricultural production of the isthmus is for export, a fraction that has held fairly steady since the 1920s; exports of unprocessed foodstuffs continue to provide more than one-half of isthmian trade. Five commodities dominate the commercial agricultural production of the isthmus: coffee, bananas, sugar, cotton, and beef.

In the late 1980s about 13 percent of Central America was classified as arable, the same proportion as in Mexico or the Southern Cone countries of South America. A further 26 percent was in pasture, and about 35 percent was under forest. Almost one-fifth of the arable land was cultivated; the rest was in temporary meadows or left fallow. More than one-third of the cultivated land was irrigated.

Coffee is the oldest and most valuable export staple; production is almost on a par with Colombia and contributes about a tenth of the world's total. During the 19th century the coffee growers created not only plantations (fincas) but also the other material facilities needed to market coffee overseas. They established national banking systems and financed many of the railways, ports, and other elements of a new outward-looking infrastructure and became the political elite of the isthmus. During the 20th century their political and economic power has continued.

Most coffee is grown on the Pacific side of the isthmus. There, seasonal rains provide the moisture needed by the plant during the early part of the growing season, and the succeeding dry months allow the bean to ripen to perfection. Coffee is grown on shaded slopes between elevations of 1,000 and 5,600 feet, where the well-drained, volcanic soils are naturally fertile. Coffee picking makes large demands on temporary labour, and about a million pickers help bring in the harvest each year; government policies have helped ensure the requisite supply of cheap labour. Traditionally profits have been high, taxes on production low, and the rewards concentrated in a small number of hands: between 80 and 90 percent of production is controlled by the top 5 percent of growers. There are marked differences in coffee production costs and yields; yields on the Meseta Central in Costa Rica are the highest, and those on well-managed estates in El Salvador and Guatemala are six times what is typical of Honduras or Nicaragua.

Banana cultivation

Bananas have flourished best in Central America in the tropical warmth and persistent moisture of the coastal lowlands of the Caribbean. The first commercial banana plantations were begun to help pay the construction costs of the Northern Railway, built to carry Costa Rican coffee to the Caribbean port of Limón. Today Caribbean banana plantations are most intensively developed in the lower river valleys of northern Honduras and neighbouring Guatemala. In the 1930s, when banana infestation devastated the southern Caribbean, new plantations were established on the Pacific coast of Costa Rica and in western Panama.

Bananas, unlike coffee, are a perishable commodity, harvested throughout the year, and they need fast, frequent, and cheap shipping facilities for marketing. Thus, heavy capital investment in ports, railways, drainage, and the like was a prerequisite for success. Large numbers of permanent plantation labourers also had to be imported for the early Caribbean plantations, and housing and social services provided; most of the new workers were English-speaking West Indians. Major U.S. companies with the available capital and technology have dominated the banana economy from its beginning.

Sugarcane was introduced to Central America by the Spanish soon after the conquest and was grown in the foothills behind the Pacific Coast from Guatemala to Nicaragua. It is still mainly grown there, taking advantage of a climate that is warm and moist during the growing season but dry just before harvest time. Sugarcane is raised on both conventional plantations and peasant plots. Central American sugar is sold chiefly in domestic markets.

Large-scale cotton growing has been a feature of Central American agriculture only since the construction of roads along the Pacific littoral in the early 1950s opened up virgin land with rich alluvial and volcanic soils and a climate well suited to cotton. The international attack on malaria helped make these lands habitable, modern insecticides allowed man to harvest a crop, and the overseas demand for cotton attracted entrepreneurs and capital. The lakes region of Nicaragua, in particular, became a focus for large-scale mechanized cotton farms. Land was revalued and subsistence farmers dispossessed of land over which they had only traditional rights. The increased use of fertilizer, as soils needed replenishment, and insecticides, as immunities built up, emphasized the growing capital-intensive (and ecologically damaging) nature of the new machine-based agriculture. In contrast to coffee and banana growing, the demand for wage labour in cotton production was modest and seasonal. The cotton boom generally benefited those with access to capital, and it spawned cotton gins, vegetable oil factories, textile mills, and agricultural suppliers. The example of cotton led to a spread of modern agricultural practices with other crops, including rice and sugar, on the Pacific lowlands.

The modernization of Central America's cattle ranching economy began in the late 1950s with the opening in Managua of the first beef-packing station approved by the U.S. Department of Agriculture; 20 years later there were more than two dozen such plants in Central America. It was a development promoted by U.S.-backed aid agencies and directed toward exploiting the huge demand for lower-grade beef in the United States. As with bananas, much of the private capital and expertise needed to introduce improved strains of cattle and grasses, to build efficient slaughterhouses, to provide refrigerated transport facilities, and to research the market was provided by U.S. corporations. The new cattle economy prospered on land not fertile enough for cotton, bringing large areas of Pacific Central America into the export economy for the first time, at the expense of exacerbating economic and social differences. Forests had to be cleared and subsistence farmers evicted to make way for beef cattle.

U.S. input in cattle raising

The growth of towns has led to an increased production for local markets. Dairy farms and market gardens cater to the sophisticated demands of middle-class townspeople, and an increasing share of the everyday food staples grown in Central America on peasant smallholdings finds its way to urban markets. Although the majority of subsistence farmers now produce cash crops, the agricultural systems used are those that have provided a livelihood throughout the centuries. Thus, in the densely populated altiplano of western Guatemala, for example, virtually every cultivatable acre remains under corn (maize), the staff of life of the American Indians. The great variety of corn grown, sometimes planted with traditional beans and squash, and the tiny size of hand-tilled individual holdings, sometimes on quite steep slopes, creates a landscape very different from that of the plowed lands of commercial monoculture elsewhere in the isthmus. In drier, but equally densely populated, El Salvador sorghum is often grown with corn, partly to provide an insurance for the largely mestizo farmers against a poor corn harvest. Both sets of farmers, American Indian and mestizo, increasingly have come to supplement their incomes by looking for temporary work in the towns or in the coffee plantations, in the cane fields, or on the cotton farms.

On the emptier Caribbean side of the isthmus different subsistence agricultural traditions survive; cassava is the main staple, supplemented by sorghum, plantains, and rice, and slash-and-burn agriculture persists in Pacific Panama, eastern Nicaragua, and central and southern Honduras. However, this extensive form of cultivation has become less sustainable as fallow periods are shortened in response to rising demographic pressures and the forest soils lose their ability to recuperate.

Forest products

Although rain forest covers much of the Caribbean lowland, forests have economic significance only in Belize and Honduras. Until about 1960 half the value of Belizean exports was supplied by the forests. Logwood and mahogany had drawn British timber companies to Belize (and the Mosquito Coast) in the 19th century, and during the 20th Caribbean pine became worth cutting. But changing taste, alternative cheaper sources, and poor management have restricted production. Central America's commercial timber now comes mainly from coniferous forests in interior Honduras and broad-leaved forests in Costa Rica.

Fishing makes only a modest contribution to the Central American economy, although shellfish middens testify to its importance in pre-Columbian diets, and inshore fishing remains important for some coastal American Indians today. It is an underexploited resource except, perhaps, in Panama, where the completion of a new port and fishing terminal at Vacamonte in 1979 has turned fish, mainly shrimps, into an important export. Exports of fresh and frozen lobster and conch from Belize to the United States have been of local importance since 1960.

Industry. The search for placer gold in the rivers of Caribbean Central America spurred early Spanish exploration but produced little permanent settlement. In contrast, the discovery of veins of silver in interior Honduras led to sustained production and a strong mining industry. Bullion was the mainstay of the Honduran economy during colonial times and the first century of independence. In the late 1800s silver and gold accounted for three-fourths of its exports, and the mining town of Tegucigalpa was chosen as the national capital.

In the 20th century lead and zinc have become significant minerals in Honduras, but the overall importance of mining there is now modest; it is even less important in the other Central American countries. The most important mineral prospect is probably the copper deposit at Mount Colorado, Panama, one of the largest of its type in the world; its development, however, is dependent upon relatively high copper prices. Most current interest has reverted to gold, both the discovery of new deposits and the reworking of old ones.

Central America is largely reliant upon imported oil. Both Belize and Guatemala have hopes that the rich southern Mexican oil fields will prove to extend into their territories. Half the undeveloped reserves of hydroelectricity are in Costa Rica.

Manufacturing is hampered by limited raw materials, few sources of energy, and the restricted size of local markets. Between one-fifth and one-fourth of the growth in manufacturing between 1964 and 1974—visible as large factories making such products as paints, detergents, tires, paper, fertilizers, and insecticides serving the member states—was due to CACM legislation. By 1980 domestic production was meeting about two-thirds of the internal demand for manufactured goods. Since 1980 protectionist measures and cooperation have been much reduced, and economic recession has both increased the costs of imported raw materials and decreased local demand. The real value added by manufacturing in Central America peaked between 1978 and 1981 and remained below that level for the next decade. Manufacturing plays a lesser part in the life of the isthmus than it does elsewhere in most of Latin America, accounting for only one-sixth of the isthmus's GDP.

The role of manufacturing

Much of the industrial employment is in cottage industries, and only a minority works in establishments employing more than 10 people; artisans outnumber factory workers. The dominant activities are those that require relatively little capital to reach an efficient level of production. The food, beverage, and tobacco industries account for about half the value added in manufacturing and more than one-third of the employment; textile, clothing, and shoe manufacturing account for a further fifth of the value and a fourth of the labour force. The woodworking industries, including furniture and leather making and printing, are important traditional activities. In contrast, the newer industries support relatively little employment. The chemical industry is the most important of these, responding to the rising agricultural demand for fertilizers and insecticides until it was hit by the 1980s recession; oil refineries, seemingly promoted as necessary symbols of nationhood, have proliferated.

Two-thirds of Central America's industrial employment is in the 10 most populous towns, and Guatemala City is home to one-fifth of all the industrial establishments in the isthmus and has one-fifth of the total employment in manufacturing. Much of the large urban employment in providing services is unrecorded. Probably between 40 and 50 percent of the working population of the larger cities of Central America is in occupations for which there is no official record. Many in this sector are self-employed artisans, outworkers (taking in work from the formal sector), vendors, and market traders intent on making an honest living. Some are small-scale criminals, petty thieves, or prostitutes who live outside the law but according to widely understood and accepted rules; a few may be part of the larger criminal economy of the international drug trade. Panama has achieved some notoriety as a funnel for channeling South American cocaine to North America; Belize has been a conduit as well as a grower of marijuana for the U.S. market.

Tourism generates an income equivalent to about 10 percent of Central America's export trade. This is well below the potential of the industry because of widespread civil and military unrest. For example, in Guatemala the dramatic scenic attractions of the volcanic highlands and lakes, the rich archaeological remains at Tikal and elsewhere, and the ethnic appeals of the American Indian communities are negated by the activities of guerrillas; tourists have also come to be deterred from visiting Nicaragua and El Salvador, and even the relatively stable Costa Rica has suffered some tourist loss. In contrast, Panama has long had the largest income from nonresidents in Central America—a third of the receipts coming from U.S. military personnel stationed in the canal area.

Import-export balance

Trade. An important economic role has been played by trade in all Central American countries since independence, allowing a limited number of primary commodities to be exchanged for manufactured and other goods produced more cheaply abroad. In general, the value of imports has exceeded that of exports except in years of exceptional commodity prices. Coffee supplies between one-fourth and one-third of Central America's exports and bananas up to one-sixth, followed by sugar, beef, and cotton. The degree of importance of trade to individual countries varies significantly: Honduras and Costa Rica are highly dependent, El Salvador moderately dependent, and Nicaragua and Guatemala less so. Panama is in a separate category because of its own huge imports of crude oil. Half of the export earnings of Costa Rica and El Salvador during the 1970s and '80s came from coffee, as did more than one-third of Guatemala's. Bananas provide between one-fourth of Panama's and one-third of Honduras' income and about one-fourth of Panama's and of Costa Rica's. Sugar makes up one-third of the value of Belizean exports and cotton one-fourth of Nicaragua's.

The United States and Canada are the destinations of one-half of Central America's exports, and the countries of western Europe, notably Germany, the other half. Regional trade within the countries of CACM reached record levels in 1980, but political instability and civil strife combined with disinvestment, capital flight, reduced domestic demand, and the erection of national protectionist barriers to imports sharply reversed this trend. Ninety percent of regional transactions are between Guatemala, El Salvador, and Costa Rica. The United States supplies about 40 percent of the isthmus's imported goods, two-thirds of which are manufactured items. Nicaragua has been exceptional in relying heavily upon the Soviet Union and other countries of the Eastern bloc as trading partners following the trade embargo imposed by the United States in 1985. Nicaragua ran the largest trade deficit in the isthmus in the late 1980s, with imports running three to four times ahead of exports in value.

Transportation. The major transport artery of Central America is the Pan-American Highway. It runs almost the entire length of the isthmus, from Cuauhtémoc on the Mexican-Guatemalan border to Chepo in the foothills of Darién. The highway became drivable in the mid-1960s and is now paved, although it is prone to landslides in the wet mountainous sections. Engineering obstacles and the lack of capital to surmount them, as well as political differences, have stood in the way of completion of the final link (across the Darién Gap) to the South American road network. New roads have made the Pacific coastal lowlands accessible since 1950, and a limited number of others have helped incorporate the Caribbean lowlands into the national life of the isthmus. By the late 1980s, however, only about 16 percent of all roads were paved. The Costa Rica and El Salvador road networks are four times as dense as those of the other countries.

Disuse of railroads

There are less than 2,400 miles of railway track in Central America, much of it not used. The systems are fragmented, and many give preference to particular customers (for example, the banana railways of northern Honduras) rather than the general public. Most railways are state-owned, and they largely suffer from neglect and a lack of investment. Many have been supplanted by roads, and none is of major importance.

Most ports are small and, like the railways, were built for particular users. The Panama Canal is the major seaway of the region, but, with the completion of the transisthmian pipeline in 1982 and the diversion of crude-oil shipments from the canal, the number of ships daily transitting it has fallen and revenue has shrunk. In 1986 a commission was set up to consider the possibility of building a new canal (perhaps a sea-level cut) to supplement the existing one.

There are some 30 airports in Central America offering scheduled services and many other public, military, and private landing strips, some reputedly used for smuggling contraband, particularly narcotics. (D.J.Fo.)

History

PRE-COLUMBIAN CENTRAL AMERICA

Central America, an archaeological bridge connecting North and South America, was, before the arrival of the Europeans, home to various nomadic and sedentary cultures. Maya civilization occupied much of the northwestern part of the isthmus, from Chiapas and Yucatán, now part of southern Mexico, through Guatemala, Honduras, Belize, El Salvador, and into Nicaragua. Although the Maya were the most advanced pre-Columbian civilization in the hemisphere, they were never unified. Unlike the Aztec or Inca empires, their autonomous city-states remained independent, presaging the political fragmentation that would characterize Central America to the present day. What unity existed was cultural, rather than political. In addition to the Maya, other American Indian nations occupied Central America, bringing influences from both North and South America.

There is scant evidence of human life in Central America before 8000 BC. Primitive human habitation in the region before that date is likely, perhaps as early as 40,000 BC, but civilized society did not emerge until the 2nd millennium BC. Between 4000 and 1000 BC, people of the region made the transition from hunting and foraging to plant cultivation. Pottery in the Parita Bay region of Panama, dating from about 2130 BC, reflected South American cultural influence, which eventually reached as far north as Guatemala and Chiapas. Mexican influence in Chiapas dates from at least 1500 BC, and thereafter it extended as far south as Nicaragua and Costa Rica. Central America thus became a meeting ground for Mesoamerican, South American, and Caribbean peoples. After 1000 BC, organized, sedentary farming communities became numerous, and active commerce and communications developed among them. Although manioc and other tubers were important to Indians of the Caribbean coast, corn (maize) was the most important staple food of most Central Americans, accompanied by a wide variety of beans, squash, and other vegetables and fruits.

Mexican influence

The Olmec culture and other Mexican influences substantially affected the development of Maya civilization, while central Mexican Nahuatl influence challenged the Maya and stretched along the Pacific coast, notable especially among the Pipil of El Salvador and the Chorotega and Nicarao of Nicaragua. In Panama and Costa Rica, South American Chibcha influence was prevalent, while Caribbean cultural patterns penetrated the coastal plain from Panama to Honduras. Fugitives from the European conquistadores in the Caribbean increased this influence considerably at the close of the 15th century. The Miskito, Sumo, Rama, and other tribes on the Nicaraguan and Honduran Caribbean shores have survived to the present.

Emergence of the Maya. After 500 BC an advanced Mayan civilization emerged in the highlands of Guatemala and El Salvador. A large population developed at the city of Kaminaljuyú, on the outskirts of present-day Guatemala City, and other Mayan cities arose from Chiapas southeastward to Nicaragua. Early in the 1st millennium AD, Classic Maya civilization arose in the Petén region of northern Guatemala, as well as in Belize and Honduras; this probably was the result of migration from the highlands, although Pre-Classic Mayan cultures had also developed in this lowland region contemporaneously with such highland centres as Kaminaljuyú. Large new cities arose at Uaxactún, Tikal, Quiriguá, Copán, and elsewhere, characterized by great stone temples, pyramids, and markets. Although the highlands remained heavily populated, these lowland cities became the centres for a higher civilization. Overcrowding, famine, climatic change, or major geologic catastrophes may have contributed to migration, but contact between the highland and lowland peoples continued as merchants carried cultural, economic, and social traits from one place to another. Tazumal, for example, in western El Salvador, was influenced by Copán, the Maya scientific centre in northwestern Honduras. Influence from as far away as Teotihuacán in the present-day Valley of Mexico and El Tajín in Veracruz also continued.

Mayan cultural advancements

The lowland Mayan cities were located in an area that the Europeans thought of as uninhabitable because of the hot, humid climate and the accompanying insects and diseases. But the Maya developed the highest culture of pre-Columbian America, which reached its height between AD 600 and 900. Maya priests made discoveries in astronomy and mathematics comparable to similar advancements in ancient Egypt. They developed an advanced system of writing, and their hieroglyphs on stone monuments (stelae), erected throughout the Maya lands, recorded their history. Brilliant Mayan artistic and scientific achievements in ceramics, sculpture, weaving, and painting, some of which were more advanced than European accomplishments of the same era, all showed remarkable artistic sensitivity. They developed an accurate calendar and complex systems of agricultural and water management. The Maya could not match their European counterparts in technology, however. Their architectural works were also less impressive than those of ancient and medieval European civilizations. The Maya were unable to develop energy-saving machines, the wheel being employed on children's toys but not adapted to any practical application. Animals were not domesticated, and all work was done by human hand labour directed by an elite group.

Decline of the Maya. Maya civilization began to decay after AD 900, perhaps owing to stresses in the social structure, overpopulation, and deforestation. A number of important cities emerged in the late Classic period, mostly on the periphery of the classic Mayan region, in present-day Chiapas or Yucatán. The peoples of these cities, much influenced by invaders from central Mexico, built striking stone architectural monuments, but their scientific and artistic achievements were less remarkable than the earlier advances of the Classic period. Their economies remained underdeveloped, however. In the meantime, internal civil war and intervention from central Mexico sapped Maya strength and vitality. By the time of the Spanish conquest, Maya civilization was thoroughly in decline, yet they resisted subjugation longer than either the Aztec of Mexico or the Inca of Peru. Returning to northern Guatemala, where they established the city of Tayasal as a place of refuge, some Maya maintained their autonomy until 1697. Not far away, the Lacandón Maya defied pacification throughout the Hispanic period, resisting from remote jungle and mountain refuges along the Usumacinta River. Disease and the social disruption brought with the Spanish conquest annihilated a large part of the native population during the 16th century. Although estimates of the pre-Columbian population of Central America vary widely, it is generally agreed that the region's population did not again reach its pre-Columbian level until the 20th century.

THE SPANISH CONQUEST

Early Spanish claims

Rodrigo de Bastidas was first to establish Spain's claim to the isthmus, sailing along the Darién coast in March 1501, but he made no settlement. A year later Christopher Columbus, on his fourth voyage, sailed along the Caribbean coast from the Bay of Honduras to Panama, accumulating much information and a little gold, but again making no settlement. Other navigators from Spain followed, some seizing natives as slaves, and in 1509 Fernando V, the king of Spain, granted concessions for colonization of the region to Alonso de Ojeda and Diego de Nicuesa. Both suffered staggering losses from disease, shipwrecks, and hostile natives. Remnants of these expeditions, under the leadership of a stowaway, Vasco Núñez de Balboa, who had earlier been with Bastidas, survived at Santa María la Antigua del Darién, on the Gulf of Urubá near the present-day Colombia-Panama border. Balboa turned the survivors into a disciplined and productive colony in 1510. Crossing the isthmus, Balboa discovered the "South Sea" (Pacific Ocean) in 1513 and claimed for Spain all the lands it touched. Balboa cultivated good Indian relations, made extensive explorations, and found enough gold and pearls to make Castilla del Oro, as it was called, the first profitable colony in the New World. The explorations took their toll on the Indians of the region, however, many of whom were wiped out by European diseases.

Appointment of Pedrarias. The king relieved Balboa with a trusted general, Pedro Arias de Ávila (known as Pedrarias Dávila), although he allowed Balboa to continue his explorations on the Pacific coast. Pedrarias, however, distrusted the ambitious Balboa and, accusing him of treason, had him beheaded in 1517. Pedrarias expanded the colony but was responsible for enslaving and murdering the Indian population, despite royal orders for more humane treatment. In 1519 he established Panama City on the Pacific coast and moved the capital there in 1524, abandoning the hot, humid Darién.

Pedrarias sent a kinsman, Gil González Dávila, to explore northward, and he found civilization on the shores of Lake Nicaragua. The jealous Pedrarias forced him to flee to Santo Domingo before a Spanish colony could be planted, however, and instead sent Francisco Hernández de Córdoba in 1524, who established Granada on Lake Nicaragua and León on Lake Managua. But when Córdoba attempted to set up a kingdom independent of Panama, Pedrarias came to Nicaragua himself and put Córdoba to death after a year of civil war.

While Pedrarias and Córdoba conquered lower Central America, the conqueror of Mexico, Hernán Cortés, looked southward. In 1524 he sent Cristóbal de Olid by sea to Honduras and Pedro de Alvarado overland to conquer Guatemala. Olid founded the port of Triunfo de la Cruz but immediately declared himself independent of Cortés, a common practice among the conquistadores. Accompanied by a large force of Indian warriors from central Mexico and preceded by a smallpox epidemic, Alvarado faced little opposition until he reached Guatemala. There he allied with the Cakchiquel Maya against the rival Quiché. He allegedly defeated the Quiché chief, Tecúm-Umán, in hand-to-hand combat at Xelajú, near present-day Quezaltenango. Alvarado went on to conquer the Pipil of El Salvador in the same year, but a bloody rebellion by the Cakchiquel took four more years to quell.

Internal struggle of the conquistador leaders

In Honduras a three-way struggle developed between the forces of Pedrarias, Cortés, and González, who had returned to Central America to press his claim to Nicaragua. The discovery of gold in Honduras made the struggle more intense. Cortés first sent Francisco de Las Casas to relieve the rebellious Olid, but then marched to Honduras himself to reprimand Olid. Before he arrived, however, Las Casas and González had united against Olid and put him to death. Cortés' difficult trip to Honduras thus turned out to be unnecessary, but before leaving he consolidated his control of the Honduran coast with the establishment of Puerto Natividad (renamed Puerto Cortés in 1869). The loyal Alvarado consolidated Cortés' control over Honduras as well as Guatemala and El Salvador, confronting the forces of Pedrarias, with whom rivalry continued for years.

Alvarado went on to participate in the conquests of Peru and of northern Mexico, while retaining his governorship of Guatemala. Uniquely, upon his death in 1541 in Jalisco, his widow, Beatriz de la Cueva, succeeded him as governor of Guatemala, chosen by leading officials in the Guatemalan capital upon news of Alvarado's death. The rule of Doña Beatriz, however, lasted but two days, for she died when a massive flood and mud slide destroyed the city on Sept. 10, 1541. Construction of a new capital, Santiago de los Caballeros de Guatemala (present-day Antigua, Guat.), began a few miles away in 1543.

Further conquests of the Indians. Indian resistance delayed the conquest of Costa Rica until 1561, when Juan de Cavallón led a successful colonization expedition there. Although none of his settlements in the Nicoya Bay region survived, he and his men began the permanent Spanish occupation of Costa Rica. A year later Juan Vásquez de Coronado took over as governor of Nicaragua and Costa Rica, and in 1564 he established Cartago as the seat of government in the central valley of Costa Rica, where a small but industrious population developed.

Spanish domination of Central America was achieved by relatively few Spanish military forces, but at a cost of millions of Indian lives. Remote areas, however, especially in northern Guatemala and along the Caribbean coast, remained outside Spanish control throughout the colonial period, eventually allowing Great Britain to colonize Belize and the Mosquito Coast of Nicaragua.

Reforms of Bartolomé de Las Casas

Yet the conquest was not entirely military. A Dominican friar, Bartolomé de Las Casas, made a notable effort to ameliorate treatment of the Indians in Central America. The brutality of the Spanish conquest had repelled Las Casas in the Caribbean. After his Dominican monasteries in Nicaragua and Guatemala failed to bring better treatment to the Indians, he went to a province of northern Guatemala to pacify it without military force. His experiment in this province, which he called the Verapaz, was only partially successful, but it served as the basis for his arguments to the Spanish crown against abuse of the Indians. The resulting New Laws of 1542 began the suppression of the encomienda system of exploitation of Indian labour.

THE HABSBURG PERIOD (1524–1700)

Unification of the isthmus. Political jurisdiction over Central America under Spanish rule evolved slowly because of the rivalries among conquistadores. These rivalries led to violence and civil war among the Spaniards in the early years of colonial rule and retarded the unification of Central America. Decentralization of authority characterized the Habsburg period, despite royal efforts to maintain close control through its agents. Municipal councils (ayuntamientos) were the most important governing units in the early days. By 1530 Guatemala, Nicaragua, Honduras, Chiapas, and Panama all functioned under separate royal orders, but the death of Pedrarias in 1531 and the prestige of Alvarado contributed to the unification of the isthmus thereafter. In 1535 the establishment of the Viceroyalty of New Spain at Mexico City included the northern portion of Central America, but the establishment of an audiencia at Panama in the same year continued the confusion over jurisdiction in Nicaragua. In 1543 Spain unified the entire isthmus from Tabasco and Yucatán to Panama as the Audiencia de los Confines, with its capital centrally located in Honduras in 1544 at the gold-mining boomtown of Gracias. The gold soon gave out, however, and the town was otherwise isolated and remote. Responding to protests from Panama City and Santiago de Guatemala, in 1548 the crown moved the capital back to Santiago. Philip II moved the capital to Panama from 1563 to 1567, but finally, after removing Panama, Tabasco, and Yucatán from its jurisdiction, he restored Santiago as the capital of the Kingdom of Guatemala, stretching from Chiapas to Costa Rica.

Development of Guatemalan dominance

Rise of the Creole elite. Santiago grew to become the third largest city in the hemisphere (after Mexico City and Lima), and established itself as a bureaucratic, ecclesiastical, and commercial metropolis. The Creole elite that emerged there favoured the development of Guatemala over the other provinces of the kingdom, causing provincial resentment and contributing to the eventual political fragmentation of the isthmus. Royal authority over the region was exercised by an audiencia, presided over by the royally appointed president, who also carried the titles of governor and captain general. Creole power was centred in the municipal council of Santiago and in their control of the land and labour of the Indians through the institutions of encomienda and repartimiento, which involved the collection of tribute and forced labour. While Creole elites developed in each province as well, and their bloodlines often were connected to the Guatemalan elite, the provincial elites were less wealthy and powerful than their Guatemalan counterparts. The Roman Catholic church participated closely with the state in developing colonial Central America, but it was strongest near the seats of authority and weakest in remote areas such as Costa Rica.

Colonial economy and society. Spain encouraged the mining of precious metals, but Central American deposits were thin, and agriculture came to dominate the economy of the colony. Cacao, mostly grown on the Pacific coast, was the principal export of the 16th century, but in the 17th century it declined because of competition from areas with better access to markets. Indigo eventually replaced it as the principal Central American export. Yet most people were involved only in subsistence agriculture, with large haciendas, in feudal style, raising cattle and grains for the local population. In the 17th century especially, as European rivals and Caribbean-based buccaneers raided Spanish commerce and shore settlements, the Kingdom of Guatemala withdrew into a self-sufficient, feudallike existence.

The colonial social structure that became entrenched during the Habsburg period thus comprised two small upper classes, one representing official Spanish administrative and ecclesiastical authority and the other the Creole landholding elite, with a large mass of Indian or mestizo rural workers tied to the land. There were also small numbers of African slaves brought during the colonial period. In the cities there were small middle sectors of artisans, provisioners, and wage labourers, but they did not constitute a true middle class. Nor were the professionals of the cities a middle class, for they were more clearly associated with one or the other of the two upper classes.

Importance of Panama

In Panama the river and mule trail across the isthmus was the principal economic resource for the commercial and bureaucratic elite that developed there. As the link between Europe and the rich mines of Peru, Panama was of strategic importance and received considerable military protection against attacks from marauding buccaneers such as the Welshman Henry Morgan, whose destruction of Panama City caused the Spanish to move the city several miles and rebuild it in 1671. Politically, Panama was tied to the Viceroyalty of Peru until the 18th century, when it was included in the new Viceroyalty of New Granada, with its capital at Bogotá. Panama's importance as a commercial and slave-trading centre justified its having its own audiencia from 1753.

THE BOURBON CENTURY (1701–1808)

Accession to the Spanish throne by the Bourbon Philip V at the beginning of the 18th century plunged the empire into a costly war, opening a century in which war and international rivalry seriously altered Central American history. The close relations of the Bourbons to France and the penetration of Central America by English traders, especially via their settlements at Belize and the Mosquito Coast, brought significant foreign commercial, administrative, military, and ideological influences. While the Spanish may have diffused the Enlightenment, it nonetheless contributed to vital changes in Creole thinking. Bourbon policy, especially after 1750, was directly responsible for much of the change, as it centralized authority and reasserted the royal control that had diminished during the previous century, while beginning to limit the political and economic power of the clergy. It also built up the military and promoted agricultural exports, especially of Salvadoran indigo, but also of Costa Rican cacao and tobacco. The Bourbon emphasis on exports began a trend in Central American economic history that would continue to the present. Indeed, the Bourbon reforms not only laid the foundation for much of Central America's political and economic development in the 19th century but they also heightened the strong regionalism on the isthmus, as provincial elites resisted the growing power of the Guatemalan mercantile and bureaucratic establishment.

An earthquake destroyed Santiago de Guatemala in 1773, causing the capital to be moved to the present site of

Guatemala City in 1776. As the century closed, the growing preference of the crown for appointment of Spaniards contributed to Creole resentment against royal policy. At the same time, there emerged in the Guatemalan capital a group of progressive Central Americans who promoted liberal economic and political ideas, especially through the publication of the *Gazeta de Guatemala* beginning in 1793 and the establishment of an economic society in 1795.

INDEPENDENCE (1808–23)

Disruption of European wars

Despite revitalization of the colonial economy and of Spanish military strength under the Bourbons, the French Revolution and subsequent Napoleonic Wars brought disintegration to Spain's empire. The Kingdom of Guatemala suffered hard times resulting from the disruption of Spanish shipping in the war. Combined with locust plagues and competition from other producing areas, this caused a decline in indigo exports during the first two decades of the 19th century. The French invasion of Spain in 1808 increased the difficulties by adding burdensome taxes and demands for "patriotic donations" to support the resistance against the French; nevertheless, the kingdom remained loyal to the Spanish government at Cádiz during these difficult years. That government, ruling in the name of the captive Prince Ferdinand, made major reforms in an effort to maintain colonial loyalty and support. The Cádiz Constitution of 1812 provided for colonial representation in the Spanish parliament and elections for municipal and provincial offices. These innovations triggered intense political activity, greatly increasing the importance of municipal and provincial councils and laying the foundations for Creole democracy.

A strong captain general, José de Bustamante y Guerra (1811–18), and Creole fear of Indian uprisings were factors that prevented Central Americans from seizing power as had been done in South America. The government easily put down such attempts in San Salvador, Nicaragua, and Guatemala. In 1814, after the defeat of Napoleon, Ferdinand VII promptly annulled the 1812 constitution. This ungrateful act caused Creole opposition to Spanish rule in Central America to mount, especially against the repressive rule of Bustamante. The restoration of the constitution in 1820 once more permitted popular political activity in Central American towns and led to the emergence of factions that would form the basis of the Liberal and Conservative parties destined to dominate Central America for the following century.

A council of notables in Guatemala City accepted the independence plan of the Mexican Creole Agustín Iturbide on Sept. 15, 1821, but there were wide differences among the municipalities on the next step. Some favoured independence from Mexico as well as from Spain, and some of the provinces wanted independence from Guatemala as well. This divisive action by the municipalities was a product of their newly acquired vitality under the constitution, but it also reflected their resentment against centralized authority in Guatemala. Conservatives in Guatemala succeeded in annexing the kingdom to Iturbide's Mexican empire, but this led immediately to civil war, as San Salvador and Granada refused to accept the decision. Mexican and Guatemalan troops subdued San Salvador after a long siege, but in the meantime Iturbide's empire collapsed and was succeeded by a liberal republic that allowed Central America to go its own way.

THE UNITED PROVINCES (1823–40)

The assembly of July 1, 1823

A Liberal-dominated assembly elected from all the provinces convened in Guatemala, and on July 1, 1823, it declared the independence of the former kingdom under the name the United Provinces of Central America. In 1824 it adopted the Constitution of the Federal Republic of Central America, a document similar to the Spanish Constitution of 1812, providing for a federation of Guatemala, El Salvador, Honduras, Nicaragua, and Costa Rica. Chiapas had elected to stay with Mexico, and Panama had become part of the Republic of Colombia in 1821.

The first election. The 1824 constitution provided for a single-house legislature and reserved considerable autonomy to the states, yet it offered an adequate framework for a strong union. Political difficulties from the outset and the failure of federal leaders to enforce the constitutional provisions led to its disintegration. The provincial jealousies and ideological differences that had emerged in the late colonial period had already sown the seeds of Central American disunion. The first presidential election, in 1825, was disputed and began a pattern of civil war and bad faith. Manuel José Arce, a liberal Salvadoran army officer, won that election over a moderate Honduran attorney and prominent intellectual, José Cecilio del Valle, despite the appearance that del Valle had more popular votes. The intrigue connected with the electoral process alienated not only Conservatives supporting del Valle but also extreme Liberals who accused Arce of selling out to Conservatives in the congress in order to gain their votes. Arce did, in fact, ally himself with Conservative interests in Guatemala City. When he deposed the Liberal Guatemalan state governor, Juan Barrundia, and replaced him with a staunch Conservative, the Salvadoran state government rebelled, touching off a civil war from 1826 to 1829. The bloody struggle established animosities between Conservatives and Liberals throughout the federation that would last well beyond the brief life of the United Provinces. The Honduran general Francisco Morazán led the Liberals to victory in 1829. Under his presidency during the decade following, the Liberals exiled leading Conservatives, including the archbishop and other clergy, and instituted sweeping anticlerical, economic, social, educational, and judicial reforms. Morazán also moved the federal capital to San Salvador in 1834.

Rebellion of Anastasio Aquino

Morazán's presidency. Resistance to Liberal policies against Indian interests surfaced in El Salvador in the rebellion of the Indian leader Anastasio Aquino beginning in 1833, but Morazán succeeded in repressing this insurrection. In the presidential election of 1834, the opposition candidate José Cecilio del Valle defeated Morazán, but he died before taking office, leaving Morazán as president. In Guatemala, opposition to the Liberal policies of Governor Mariano Gálvez, including anticlericalism, encouragement of foreign immigration, land grants, judicial reform, and a general head tax, combined with panic caused by a cholera epidemic, led to peasant revolts beginning in 1837. Behind the charismatic leadership of Rafael Carrera, the peasants not only toppled Gálvez but also sharply divided the Liberals in Guatemala and allowed the Conservatives to gain control. Taking advantage of these problems, the western departments of Guatemala, under Liberal leadership, seceded and formed a sixth state, called Los Altos. Carrera quickly reconquered these departments in January 1840, however, and when Morazán brought federal troops into the conflict, Carrera defeated him soundly at Guatemala City in March 1840. The federation was already in disarray, as Nicaragua, Honduras, and Costa Rica had seceded in 1838. Morazán fled to Panama.

FORMATION OF THE REPUBLICS (C. 1840–C. 1870)

Rafael Carrera quickly dismantled the Liberal program in Guatemala and supported Conservative caudillos in other Central American states. Although many gave lip service to reunification, all attempts failed, and Conservative rulers in all the states opposed reunification.

Morazán returned in 1842 and seized power in Costa Rica, seeking to make it a base for restoration of the federation. He found little support for this and was himself ousted by Costa Rican Conservatives and executed in San José on Sept. 15, 1842.

In 1847 Guatemala declared itself a sovereign republic, quickly followed by Costa Rica in 1848, and eventually by the other regional states. The alliance of Nicaraguan Liberals with the American filibuster William Walker in 1855 caused Central Americans from all five states to unite against Walker, who made himself president of Nicaragua in 1856. In what became known as the "National War," this united army defeated Walker in 1857. Yet attempts to turn this united effort into a new federal union gained little support from the Conservative elites in each state; thus, the most lasting legacy of the Conservative period was the fragmentation of the United Provinces into the five city-state republics. The middle of the century also witnessed strong British-U.S. rivalry in Central America for commercial rights and control of transisthmian transportation routes. Early 19th-century British commercial dominance later gave way to U.S. economic, diplomatic, cultural, and military dominance in the region.

THE LIBERAL PERIOD (C. 1870–C. 1945)

The death of Rafael Carrera in 1865 signaled a Liberal resurgence throughout Central America. By 1872 the Liberals had returned to power in all the states except Nicaragua, where the legacy of alliance with William Walker had so discredited the Liberals that it delayed their return to power until 1893. Liberal domination of Central America from about 1870 through the mid-20th century resulted in a completion of anticlerical reform and strong emphasis on agricultural exports as the key to national modernization. Coffee became the most important commodity promoted by Liberals, and it supported the rise of planter elites in most of the states. Banana exports, developed in the coastal regions by U.S. fruit companies (notably United, Standard Fruit and Steamship, and Cuyamel) in collaboration with the Liberals, also were important in developing the transportation and communications infrastructure and in bringing Central America more fully into the North Atlantic trading economy. Despite their Liberal political rhetoric, military dictatorships were the characteristic political institution of the period, as the planter elites depended on greater military strength to defend their interests.

Liberal agricultural export emphasis

MODERN CENTRAL AMERICA (C. 1945 TO THE PRESENT)

By the middle of the 20th century the powerful political and economic elites associated with the export-led economies promoted by the Liberal parties faced strong challenges from middle- and working-class representatives. This challenge to the elite parties took many forms, from formation of broader-based political parties to violent revolution, accounting for most of the political crises of the mid- and late 20th century. The demands for significant socioeconomic reforms brought revolts to every state, and Central American politics in the late 20th century became characterized by strong conflict between free market and Marxist development models. Also influencing political trends was the growth of Evangelical Protestantism.

Although the Liberals had traditionally favoured unification, the strength of the local elites in each of the republics prevented numerous attempts at reunification from succeeding even under Liberal rule. Modern manifestations of the continued concept of Central American nationalism and desire for unification were seen, however, in the formation of the Organization of Central American States in 1951, followed by the formation of the Central American Common Market in 1960 and the 1987 Central American peace plan, also called Esquipulas II, instigated by President Oscar Arias Sânchez of Costa Rica. The latter included plans for a Central American national parliament along lines similar to the European unification plan. While state sovereignty and the strength of the individual city-state elites remain strongly rooted in the Central American political tradition, there continues to be a strong residue of sympathy for Central American reunification. (R.L.W.)

For later developments in the history of Central America, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

For coverage of related topics in the *Macropædia* and *Micropædia,* see the *Propædia,* sections 966 and 974, and the *Index.*

THE COUNTRIES OF CENTRAL AMERICA

Belize

Belize (called Belice in neighbouring Spanish-speaking countries), a constitutional monarchy and a member of the Commonwealth, is located on the east coast of Central America. After El Salvador, it is the smallest political unit on the mainland of the Americas. It has an area of 8,867 square miles (22,965 square kilometres). To the north and northwest, it is bounded by Mexico, to the south and west by Guatemala, and to the east by the Caribbean Sea, on which it has a 174-mile (280-kilometre) coastline. Belize City was ravaged by a hurricane in 1961, and in 1970 a new capital, called Belmopan, was built about 50 miles (80 kilometres) inland. Belize was the last British colony on the American mainland and achieved independence Sept. 21, 1981.

The name Belize is traditionally believed to have been derived from the Spanish pronunciation of the last name of Peter Wallace, a Scottish buccaneer who may have begun a settlement at the mouth of the Belize River about 1638. It is also possible, however, to trace the name to any of several Mayan words.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief and drainage.* Situated south of the Yucatán Peninsula, Belize is a land of mountains, swamps, and tropical jungle. The southern half of the country is dominated by the rugged Maya Mountains, an igneous plateau cut by erosion into hills and valleys that stretch in a southwesterly to northeasterly direction. The Cockscomb Range, a spur of the Maya Mountains, runs toward the sea, culminating in Victoria Peak (3,681 feet [1,122 metres] high), the highest point in Belize. The northern half of the country consists of limestone lowlands less than 200 feet (60 metres) above sea level, much of which are swamp. The lowlands are drained by the navigable Belize River (on which stands Belize City), the New River, and the Hondo River; the latter forms the northern frontier with Mexico. Both the New and the Hondo rivers drain into Chetumal Bay to the north. South of Belize City the coastal plain is crossed by short river valleys. About 15 miles off the coast the second largest barrier reef in the world runs parallel to the coast and is fringed by dozens of small islands called cays.

The Maya and Cockscomb mountains

Climate. Belize has a subtropical climate, with a well-marked dry season from late February to May and a wet season from June to November that is interrupted from August to September by another dry season. The mean temperature at Belize City is 74° F (23° C) in December and 84° F (27° C) in July. The mean annual rainfall increases sharply from 51 inches (1,295 millimetres) at Corozal on the northern frontier to 175 inches (4,445 millimetres) at Punta Gorda in the south. At Belize City

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A line of waves breaking against the extensive barrier reef of Belize within view of a forested offshore cay.

rainfall amounts to 74 inches (1,880 millimetres), but there are, here as elsewhere, considerable yearly variations. Trade winds blow onshore most of the year, and from September to December northerly winds bring cooler, drier air. Hurricanes are a threat from July through November.

Plant and animal life. Almost half of Belize consists of forests. There are at least 50 different forest tree species, which include mahogany, Santa Maria, cedar, and ironwood. In the north limestone soils support deciduous forest, and sapodilla and mahogany predominate. In the south the forest is taller and is evergreen. Santa Maria, rather than mahogany, flourishes on the plateau, and oak and pine grow on some of the plateau ridges. The rivers are largely bordered by swamp forests. On the southern coastal plain and inland from Belize City, open savanna (grassland) is marked by scattered oaks, pines, and palmetto palms. The coast is fringed with mangrove trees.

The abundant wildlife of Belize includes such animals as deer, jaguar, puma, tapir, American crocodile, and manatee, as well as many species of turtles, tortoises, birds, reptiles, insects, and fish.

Farm types

Settlement patterns. Only a small proportion of the land is actively utilized, and a substantial part of this is agricultural land. Nearly all the farms are less than 100 acres (40 hectares) in area, many of them *milpas* (temporary forest clearings). On most of these farms, traditional shifting agriculture is practiced, largely because of the nutrient-poor soils of the lowlands. The remaining farms or plantations are large, devoted to the raising of export crops, mainly sugarcane, citrus fruits, and bananas. In the 20th century, the government has tried to increase the production of export and domestic crops. The highlands are mostly forested and largely uninhabited.

Belize City, the former capital and most populous town, is located just above sea level and is surrounded by a mangrove swamp. Its growth has been unplanned, and it contains a mixture of old colonial structures, wooden frame buildings, and new concrete houses.

The people. Many Belizeans are of racially mixed ancestry. English-speaking people of largely African and European ancestry, who are called Creoles, predominate in the central coastal regions, while Maya Indians and mestizos (of mixed Maya and European ancestry) predominate in the more sparsely inhabited interior. Several thousand Garifuna, formerly called Black Caribs, who are descendants of the Carib Indians and Africans exiled from British colonies in the eastern Caribbean (Lesser Antilles) in the 18th century, live in communities on the south coast.

English is the official language, but most of the population also speak a Creole patois. The mestizos speak Spanish, and the Maya speak Spanish or Yucatec, Mopan, or Kekchi Maya. The Garifuna speak their own Arawakan-based language and also English or Spanish. Many Belizeans are multilingual.

Anglicans, who established the first church in the early 19th century, were soon followed by Baptist and Methodist missionaries. The Roman Catholic church was established in 1851, and more than three-fifths of the population are adherents of that religion. The next largest denominations are Anglican and Methodist, but evangelical and fundamentalist churches, many of them based in the United States, are growing rapidly.

Immigration

Migration patterns are changing the ethnic composition of the population. Groups of Mennonites migrated from Mexico and Canada after 1958 and established agricultural settlements to the north and west of Belize City. In the 1980s, an estimated 25,000 Spanish-speaking refugees, mostly from Guatemala and El Salvador, settled in Belize, while an even larger number of Belizeans, mostly Creole, migrated to the United States.

The economy. *Agriculture, forestry, and fishing.* The export of timber was for years the basis of the Belizean economy, but by 1960 the combined value of sugar and citrus exports exceeded that of timber. Chicle, used in the manufacture of chewing gum, is obtained from the sapodilla tree, and mahogany, pine, cedar, and rosewood have increased in economic importance. Furniture and power and telephone poles are the major products of the forestry industry, which includes many sawmills.

More than one-fourth of the population is engaged in agriculture. Sugarcane is grown in Corozal and Orange Walk districts, and sugar is exported to the United States and the European Economic Community. Citrus fruits (oranges and grapefruit) and bananas are grown mainly in the Stann Creek and Cayo areas, south and west of Belize City. Rice is raised on large mechanized farms in the Belize River valley, while corn (maize), roots and tubers, red kidney beans, and vegetables are raised throughout the country, mostly by small farmers. Marijuana is widely, though illegally, grown in Belize for illegal export to the United States.

The Mennonite community at Spanish Lookout has developed fresh milk and cheese production and has introduced large-scale chicken farming. Beef cattle and pigs are raised in many parts of Belize.

Fishing for lobster, shrimp, scale fish, conch, and sea turtles is conducted mainly by several cooperatives, some of which have freezing plants. Exports of seafood to the United States are substantial.

Industry. Manufacturing, accounting for about one-eighth of the gross national product (GNP), is oriented toward both the domestic and export markets. The Belize government has stressed the substitution of imports to promote industrial development. Fertilizer and animal feed plants have been opened, while the numerous sawmills, a wire and nail plant, and a roofing-material plant serve the construction and furniture-manufacturing industries. Footwear, beer, soft drinks, and cigarettes are also produced. Food-processing industries include a sugar refinery at Tower Hill, citrus-processing plants in the Stann Creek district area, beef-packing and rice-processing plants, and fish-freezing and canning factories. Garment factories utilizing imported fabric produce clothing for the export market.

Transportation. Agricultural and forest produce is usually transported by road, although rivers are still used. The road network extends west to the Guatemalan border and north to the Mexican border. All-weather roads link Belize City and Belmopan with other towns in the central and northern areas of Belize and with Punta Gorda on the southern coast.

The ports

Belize City is the main port but does not have modern facilities; vessels with more than the allowable draft must still anchor more than a mile offshore. Another port, at Commerce Bight, handles the citrus exports of the Stann Creek district; smaller ports are at Riversdale and Big Creek, and Punta Gorda handles seaborne trade with Guatemala and Honduras.

The Philip S.W. Goldson International Airport is about nine miles from Belize City; scheduled flights link it to the United States, Mexico, and other countries of Central America, and to the Caribbean through Jamaica. There are also scheduled domestic flights to Belmopan and other major towns, including Independence (near Big Creek) and San Pedro on Ambergris Cay.

Government and social conditions. The structure of government is based on the British parliamentary system, and the legal system is based upon the common law of England. The 1981 constitution provides for universal suffrage for citizens over the age of 18 and for a bicameral National Assembly composed of an elected House of Representatives with 28 members and an appointed Senate with 8 members. General elections to the House must be held within five years of the preceeding election. The governor-general, a Belizean national who represents the British crown, normally appoints as prime minister the leader of the majority party in the House and an opposition leader from the minority party. The prime minister appoints the Cabinet. Local government consists of the Belize City Council and six town boards with authority over most municipal affairs. Most villages have councils, and some Maya villages have an *alcalde* (mayor) with limited powers. A chief justice heads the Supreme Court, the nation's highest court, which is independent of the national government.

The vast majority of the population over 15 years of age is literate. Primary schooling is compulsory between the ages of 6 and 14. Most schools are government-subsidized denominational schools, chiefly Roman Catholic, and the Mennonites operate their own schools. About one-third of the secondary schools are run by the government, and the remainder are managed by the churches. About half of the primary-school graduates go on to secondary school, but only a tiny elite receive any higher education. The University College of Belize and an extramural department of the University of the West Indies in Belize City provide postsecondary education.

Health and welfare

The majority of Belizeans have access to government hospitals, clinics, and maternal, child-care, and dental facilities. As a result the people's health has improved. Infant mortality rates, in particular, have been reduced by improved water supplies, waste disposal systems, and disease control and vaccination programs. Malaria, however, remains a problem. Nurses are trained locally, but there remains a shortage of doctors and dentists, especially in the rural areas. A new social security program was created in the 1980s to provide pensions for senior citizens and to extend assistance to pregnant, sick, disabled, and unemployed workers and to the survivors of deceased insured workers.

Cultural life. The small population of Belize is culturally diverse. Most of the people are the descendants of immigrants from Europe, Africa, and India, as well as of the indigenous Maya, and of the various mixtures of these peoples. The multiplicity of languages, religions, modes of dress, cuisines, styles of music, and folklore reflects the cultural mix. While there are many ethnically distinct communities, people of different groups also mix in many social contexts: at work, in schools, and in the political parties that are not ethnically based. Though prejudices exist, there is no history of interracial or interethnic violence in Belize.

Since World War II, Belizeans have created a variety of institutions to meet their social and cultural needs, including trade unions, credit unions, cooperatives, and many other nongovernmental organizations that address the needs of health care, social services, women's rights, education, and community development.

The Baron Bliss Institute in Belize City promotes the arts and houses a small archaeological museum and the National Library Service, which has branches in the districts and operates mobile libraries throughout the country. The national archives are in Belmopan.

The government-operated Belize Radio broadcasts in Spanish and English. There is no national television station, but Belize receives U.S. channels via satellite. Most of the nation's several weekly newspapers are politically affiliated. Press censorship is minimal, but most of the journalism is of poor quality, and Belize relies heavily on external, chiefly U.S., news sources. *Belizean Studies,* a journal published three times a year by St. John's College (a secondary school), is an outlet for local research and writing, as is Cubola Productions, which publishes both fiction and nonfiction. The Society for the Promotion of Education and Research (SPEAR) publishes books and reports and a monthly newspaper on Belizean issues.

(A.E.A./O.N.B.)

For statistical data on the land and people of Belize, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Maya Indians lived in the area now known as Belize for centuries before the arrival of Europeans. The Spanish penetrated the area in the 16th and 17th centuries and tried to convert the Maya to Christianity, but with little success. It was British buccaneers and logwood cutters who finally settled on the inhospitable coast in the mid-17th century.

British treaty rights for logwood

Spain regarded the British as interlopers in their territory. By treaties signed in 1763 and 1783, Spain granted to British subjects the privilege of exploiting logwood and, after 1786, the more valuable mahogany, within specified but poorly surveyed limits. Spain retained sovereignty over the area, which Britain called a settlement, as distinct from a formal colony. While formal government was not allowed by Spain, the British settlers conducted their affairs with a public meeting and elected magistrates. Superintendents, appointed by the British government after 1786, slowly established their executive authority at the expense of the settlers' oligarchy. In 1798 the British defeated the last attempt of Spain to remove them by force, and Belize became a colony in all but name. The superintendent was instructed by the British government to assume authority over the granting of land in 1817, and he assumed the power to appoint magistrates in 1832. In 1854 a constitution formally created a Legislative Assembly of 18 members elected by a limited franchise, and the next year the Laws in Force Act validated the settlers' land titles.

Guatemala challenged the British occupation on the grounds that it had inherited Spanish interests in the area, and from time to time Mexico also asserted a claim to part of Belize. Great Britain and Guatemala appeared to have settled their differences in 1859 by a treaty that defined boundaries for Belize. The final article of the treaty, however, bound both parties to establish "the easiest communication" between Guatemala and Belize. When the communication route was not developed, Guatemala insisted that the entire treaty was thus invalidated and asserted a claim to possession of the whole territory.

Britain proclaimed Belize to be the colony of British Honduras in 1862 and a crown colony in 1871, when the Legislative Assembly was abolished. British Honduras remained subordinate to Jamaica until 1884, when it acquired a separate colonial administration under an appointed governor.

The British settlers, who called themselves Baymen, began importing African slaves in the early 18th century to cut logwood and then mahogany. Although the conditions and organization of labour in timber extraction were different from those on plantations, the system of slavery was cruel and oppressive. There were four slave revolts in Belize, and hundreds of slaves took advantage of the terrain and the freedom offered over the frontiers to escape. A flourishing transit trade with neighbouring Spanish Central America, established early in the 19th century, continued after the Spanish colonies attained independence in the 1820s, but plantations, which were forbidden by the treaties with Spain, were not developed. After emancipation in 1838, the former slaves remained tied to the logging operations by a system of wage advances and company stores that induced indebtedness and dependency. When the old economy, based on forest products and the transit trade, declined in the mid-19th century, these Creole freedmen remained very poor.

Plight of freed slaves

Beginning in the early 19th century, Garifuna immigrants settled on the southern coast of Belize. The Caste War in Yucatán between 1847 and 1853 resulted in several thousand Spanish-speaking refugees settling in northern Belize, while Maya communities were reestablished in the north and west. These immigrants introduced a variety of agricultural developments, including traditional subsistence farming and the beginning of sugar, banana, and citrus production. In the 1860s and '70s the owners of sugar estates sponsored the immigration of several hundred Chinese and East Indian labourers. In the late 19th century Mopan and Kekchi Maya, fleeing from oppression in Guatemala, established largely self-sufficient communities in southern and western Belize.

By the early 20th century, the ethnic mixture of Belize was established, but the economy was stagnant and crown colony government precluded any democratic participation. In the 1930s, the economy was hit by the depression, and Belize City was largely destroyed by a hurricane in 1931. A series of strikes and demonstrations by labourers and the unemployed gave rise to a trade union and to demands for democratization. The elective principle was reintroduced in 1936, but property, literacy, and gender qualifications severely limited the franchise. When the governor used his reserve powers to devalue the currency at the end of 1949, leaders of the trade union and the Creole middle class joined in a People's Committee to demand constitutional changes. The People's United Party (PUP) and its leader, George Price, emerged from the committee in 1950 and led the independence movement.

Belize evolved through several stages of decolonization, from universal adult suffrage in 1954 to a new constitution and internal self-government in 1964, when Price, who had been called first minister, became premier. Unrelenting Guatemalan hostility, however, impeded independence. In the 1970s, Belize internationalized its case for self-determination at the United Nations and in the Nonaligned Movement. Although the dispute between Guatemala and Great Britain remained unresolved, Belize became independent on Sept. 21, 1981. Belize was admitted to the United Nations but has been denied membership in the Organization of American States because of a rule that bars admission to states involved in territorial disputes with another member. Because the Belize Defence Force is small, a British military presence remains to deter Guatemalan aggression.

20th-century Belizean politics

The United Democratic Party (UDP), formed in 1973 and led by Manuel Esquivel, won the general election in 1984, but in 1989 the PUP won the election and Price again became prime minister, as the office was now called. Although democratic politics is well established in Belize, the struggling, dependent economy, the impact of regional civil wars, and the increasingly pervasive influence of the United States continue to make the developing nation of Belize highly vulnerable to the fluctuations and pressures of political and economic forces. (W.J.Gr./O.N.B.)

For later developments in the history of Belize, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Costa Rica

A nation of the isthmus of Central America, the Republic of Costa Rica (Spanish: República de Costa Rica) is located between Nicaragua, situated on its northern border, and Panama, on its eastern border. Costa Rica's greatest length, between the two neighbouring countries, is 288 miles (463 kilometres), and, at its narrowest, the distance between the Pacific Ocean and the Caribbean Sea is 74 miles (119 kilometres). The country's total area is approximately 19,730 square miles (51,100 square kilometres).

Costa Rica played a role in the federation of Central American states from 1823 to 1838 and is a member of the Organization of Central American States founded in 1951. Of the states that have been partners in these two enterprises, Costa Rica is the most Spanish in character and is generally regarded as having the most stable government and economy. Its well-populated heartland is devoted to harvests of coffee, its most important product, while in its outlying reaches banana cultivation is most significant. The capital is San José.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* Two mountain chains together run almost the entire length of Costa Rica. These are, in the north, the Cordillera Volcánica, noted, as the name implies, for its volcanic activity, and, in the south, the Cordillera de Talamanca. The Cordillera Volcánica may be divided into three ranges: from northwest to southeast, the Cordillera de Guanacaste, the Cordillera de Tilarán, and the Cordillera Central. The Cordillera de Talamanca is a massive granite batholith, quite different geologically from the volcanically active northern ranges. Costa Rica's highest point is Mount Chirripó, in the Talamanca system, at 12,533 feet (3,820 metres). Two of the highest peaks in the Cordillera Volcánica, Irazú (11,260 feet) and Poás (8,871 feet), have paved roads reaching to the rims of their active craters. These volcanoes overlook the densely populated upland basin called the Meseta Central (Valle Central), and they pose a serious natural hazard, as do earthquakes for most parts of the country.

Meseta Central

The Meseta Central is separated into two parts by the continental divide. The eastern part is drained by the Reventazón River to the Caribbean, and the western sector forms part of the basin of the Río Grande de Tárcoles, which flows into the Pacific. Another large structural valley, the Valle del General, lies at the base of the Cordillera de Talamanca in the southern part of the country. To the north and east of the mountainous central spine lie the Caribbean lowlands, about one-fifth of the country and less than 400 feet in altitude. The Pacific lowlands, which contain several small valleys and plains, include only about one-tenth of Costa Rica's territory.

Turning coffee beans for even drying in the Costa Rican heartland, near San José.

© Elizabeth Harris from TSW—CLICK/Chicago

Climate. Thermal convection and onshore breezes bring abundant rains to the Pacific coast in the wet season, generally May to October in the north and April to December in the south. Northeasterly trade winds on the Caribbean provide ample year-round precipitation. The higher mountain ranges have warm temperate climates, the Pacific slopes having alternatingly wet and dry seasons, while the Caribbean side has year-round rainfall.

San José's weather records report monthly averages of rainfall from well under 1 inch (25 millimetres) in February to more than 12 inches (300 millimetres) in September, with more than 70 inches (1,800 millimetres) the yearly average. Temperatures vary with altitude, San José at 3,760 feet reporting a mean of 69° F (21° C), a nearby station at 7,665 feet reporting a mean of 59° F (15° C), and another at 682 feet reporting a mean of 80° F (27° C).

Plant and animal life. Dense broad-leaved evergreen forest, which includes mahogany and tropical cedar trees, covers about one-third of the landscape. On the Talamanca range grow numerous evergreen oaks and, above the timberline, mountain scrub and grasses. The northwest, with the longest dry season, contains open deciduous forest. Palm trees are common on the Caribbean coastline, and mangroves grow on the shallow, protected shores of the Nicoya and Dulce gulfs along the Pacific.

Mammalian life is both abundant and varied and has major ties to South and North American populations. The South American species include monkeys, anteaters, and sloths; the North American, deer, wildcats, weasels, otters, coyotes, and foxes. There is a wide variety of tropical birds in the lowlands, and reptiles, such as snakes and iguanas, and frogs are common.

19th-century coffee-based expansion

Settlement patterns. Since the beginnings of European colonization, the Meseta Central has been the heartland of Costa Rica. In the 19th century, settlement slowly expanded from the core areas around Cartago and San José, the capital, into the western parts of the valley. This expansion was based on coffee production from small family farms. Such farms still prevail; more than half of Costa Rica's coffee farms are 10 acres (4 hectares) or less in size, a factor that contributes to the democratic heritage for which the country is famed.

During the 20th century, Costa Rica's settlement frontiers expanded outward rapidly from the Meseta Central to incorporate peripheral areas, until virtually all the suitable lands in the country were settled and the spread of population effectively ended.

In the Caribbean lowlands the banana industry thrived from the 1880s until the 1920s, when disease forced closure of the plantations. New, disease-resistant varieties of bananas allowed reestablishment of the Caribbean plantations in the late 1950s, thus reviving the economy. The southern Pacific coastal region was opened for settlement about 1938 by the development of banana plantations around Parrita and Golfito. After World War II, banana production declined, and the last Pacific region company-owned plantations were closed or planted in oil palm by 1985. Elsewhere in the south, habitation of the Valle del General increased rapidly following construction of the Inter-American Highway during the 1950s.

The San Carlos Plain, part of the northern lowlands, was settled mainly after 1945, when roads connected it with the Meseta Central. In the 1970s and '80s more new roads brought additional expansion of agriculture and cattle grazing to this fertile area.

The northwestern *provincia* of Guanacaste, where many people work on large cattle ranches, or haciendas, and also maintain small agricultural plots of their own, was once a part of Nicaragua and still retains a variety of Nicaraguan cultural influences. In many ways, this is the least traditionally Costa Rican part of the country.

San José is the only true metropolitan area in Costa Rica. The congested downtown contains the major stores, government buildings, and offices of many businesses. The few high-rise buildings are located in this city-centre area. Outside the downtown, San José has expanded outward to incorporate surrounding towns. The San José metropolitan area, which contains overall about half of Costa Rica's population, is a functionally integrated urban region that reaches from Alajuela on the west to Cartago on the east.

The people. *Ethnic and religious groups.* Costa Rica is noted for having the largest percentage of Spanish population in Central America. The Meseta Central, with more than half the nation's population, is the most predominantly Spanish region in both its manner of living and its ancestry. Spanish is spoken with distinctive national accents and usages. In Central America, a Costa Rican is called a Tico, for Costa Ricans replace the diminutive ending *-tito* with *-tico,* a practice known elsewhere but uncommon in Central America.

Spanish culture in the Meseta Central

The population of Guanacaste *provincia,* which makes up about 8 percent of the country's total, is a blend of colonial Spanish, Indian, and African peoples; their spoken Spanish is more like that of Nicaragua than that of the Meseta Central.

People of African ancestry live mostly in the Caribbean lowland *provincia* of Limón, which contains overall about 7 percent of Costa Rica's population. They are the descendants of workers brought from the West Indies to build railroads and raise bananas, and most of them speak both Spanish and a Jamaican style of English, the majority being descended from people who came from that island. There are also a substantial number of Chinese, many of whom are also the descendants of imported labourers.

Less than 1 percent of Costa Rica's population is Amerindian. The Bribrí and Cabécar peoples are the most numerous and inhabit valleys in the Cordillera de Talamanca. The Boruca and Térraba groups live in the hills around the Valle del General. There are also a few hundred Guatuso peoples living on the northern plains in Alajuela *provincia.* Most of Costa Rica's Indians are rapidly becoming assimilated, but those on the Caribbean side in the southern Talamanca region maintain their separate ways, including their animistic religions. Although Costa Rica's Amerindian groups are legally assigned to protected reserves, the land is infertile, and the Amerindians, most of whom survive through subsistence agriculture, are among the country's poorest people.

About 90 percent of Costa Ricans are Roman Catholics. Roman Catholicism is the official religion, and it is supported with a small part of the national budget. Most of the remaining population is Protestant, the majority of whom live in Limón province. A small Jewish congregation resides in San José.

Demographic trends. At the mid-20th century, Costa Rica's population growth rate was among the highest in the world. As general prosperity and urbanization increased, however, the population growth rate decreased, despite a drop in both the general and the infant mortality rates. Life expectancy is substantially longer than for most other Central American nations.

European immigration and customs have molded Costa Rican history and influenced its character. However, in the 1980s, immigration of largely mestizo refugees from Nicaragua and other Central American countries and from Cuba became important, representing a growing influence on the character of the country. Costa Rica has also become a mecca for retirees from the United States, tens of thousands of whom now live there.

The economy. Costa Rica is neither rich, as its name ("Rich Coast") implies, nor as poor as many of its neighbours. The country's wealth is better distributed among all social classes than elsewhere in Central America. Through the 1980s, the standard of living declined somewhat as a result of economic stagnation and inflation, and Costa Rica lost to Panama its place as the Central American country with the highest per capita gross national product.

Government controls

The government controls key utilities, including electricity, water, telephone, and port and rail facilities, and the entire population is eligible for free medical care, but private enterprise is still strong and influential in policy-making. Repeated efforts to diversify the economy have failed to reduce the traditional dependence on agriculture, but there has been some success at diminishing Costa Rica's reliance on the two traditional export crops, coffee and bananas. Despite stringent efforts to reduce spending, the Costa Rican government operates at a deficit, a condition that increases the country's already large international debt and high rate of inflation.

Resources. Costa Rica's agricultural land and climate are its most important natural resources. The country has few mineral resources. The most important are yet unexploited bauxite deposits in the General and Coto Brus valleys and copper in the Cordillera de Talamanca. There is manganese on and near the Nicoya Peninsula, gold on the Osa Peninsula and parts of the Pacific slopes, and magnetite on scattered beaches, particularly on the southern Caribbean coastline. Geologic conditions are promising for petroleum in the southern Caribbean coast, but exploration has proved disappointing. Much of Costa Rica's timber reserves were wastefully cleared to make way for pasture or cropland. The best remaining stands of tropical hardwoods are in protected parks and forest reserves. Hydroelectric power has the potential to supply domestic needs with enough surplus for export. The largest facility is the Arenal hydroelectric and irrigation project in Guanacaste, which opened in 1979.

Agriculture. More than one-fourth of Costa Rica's economically active people work in agriculture, which contributes about one-fifth of the national product. Coffee, from the highlands, and bananas, produced mainly in the Caribbean lowlands, are the most important crops, accounting for nearly half the total value of all exports. Beef, chiefly from the Pacific northwest, is the third largest export, but is declining in importance. Nontraditional exports of growing importance include ornamental plants and cut flowers and pineapple. Palm oil for domestic consumption is an important product from the southern Pacific lowlands. Costa Rica has the capacity to feed itself but dedicates a large share of its land to production of export crops. As a result, the country imports staples such as corn (maize) and beans, which it could produce, along with products such as wheat, for which the Costa Rican climate is not suitable.

Major manufactured items

Industry. Manufacturing contributes more than one-fifth of the gross national product and employs approximately one-sixth of the economically active population. Most industry is concentrated in the Meseta Central, but a few plants operate in Puntarenas and Limón. Food and beverage processing and textile, shoe, and furniture making are domestically important. The main items manufactured for export are clothing and cloth, medicines, thread, and electrical appliances. The textile industry grew in the late 20th century, largely owing to the development of plants making clothing from imported cloth for export to the United States.

A substantially larger number of Costa Ricans are employed in the service industries than in manufacturing. Commerce, finance and real estate, public administration, transport, construction, and utilities are other important branches of economic activity. Tourism has considerable potential and is developing.

Finance and trade. Costa Rica has both state-owned and private banks, and a national federation of savings and loan cooperatives supervises an extensive network of local agencies. There is a small national stock exchange. Insurance is a state monopoly controlled by the National Insurance Institute. Costa Rica is generally favourable toward foreign investment, and foreign-owned companies control a large segment of both agricultural and industrial production. Per capita national debt is among the largest in Central America.

Costa Rica imports wheat from the United States and corn and beans from its neighbours. Nonfood imports include insecticides and other chemicals, machinery, crude oil and petroleum products, and paper products. More than one-third of imports by value come from the United States, with most of the rest originating in Japan, Germany, Guatemala, and Venezuela.

Coffee is shipped in greatest quantity to the United States, western and central Europe, the Soviet Union, and Saudi Arabia. Bananas go chiefly to the United States, Italy, Germany, and Belgium. Other exports of importance are beef, textiles, fish and shrimp, sugar, pineapple, and cut flowers. About 40 percent of exports, by value, go to the United States; other countries receiving Costa Rican exports include Germany, Panama, the United Kingdom, and Guatemala.

Highway system

Transportation. The hub of Costa Rican transportation is the Meseta Central. A four-lane highway extends west from San José to beyond San Ramón. Another highway, completed in 1987, has greatly reduced distance and travel time between San José and the Caribbean lowlands. Elsewhere in the Meseta Central are narrow, often tortuous, paved routes, with few interconnections, that reach the many valley and mountain communities in the immediate area. National railroad lines link San José with Puntarenas and Limón. The Inter-American Highway (the Central American part of the Pan-American Highway) connects Costa Rica with the rest of Central America and beyond.

Limón and Puntarenas have new port facilities constructed at nearby Moín and Caldera, respectively. These modern facilities have rail access and are equipped to handle containerized cargo and, in the case of Moín, petroleum shipments. The southern Pacific port of Golfito, once an important banana shipping centre, handles little trade since the decline of banana production there. Limón is the busiest of the three ports.

Juan Santamaría, near San José, is the only international airport used by commercial jet planes. Tobías Bolaños Airport at Pavas, a suburb of San José, has a paved runway used by small aircraft. Paved runways at Limón, Golfito, and Liberia and gravel strips at a few other airports offer local service.

Administration and social conditions. *Government.* Costa Rica is governed by its constitution of November 1949, the 10th in its history. A president, two vice presidents, and a unicameral Legislative Assembly are elected at one time for a term of four years, the assembly by proportional representation. The president is not eligible for reelection. Magistrates of the Supreme Court are chosen by the assembly for eight-year terms, being then automatically continued in office unless removed by a two-thirds vote. An independent Supreme Electoral Tribunal oversees the election process.

The nation's seven *provincias* are administered by governors appointed by the president. Their importance is mainly as judicial and electoral jurisdictions, most government agencies having their own administrative organization that ignores provincial boundaries. Each *provincia* is divided into *cantones,* and each *cantón* into *distritos.* Councilmen for the *cantones* are elected locally, but budgets for all political units are approved by the national government, which controls nearly all the funds available to local governments.

Costa Rica has a stable democratic government. The fairness of national elections has been indicated by the fact that almost every four-year period since the mid-20th century has seen a change in the party winning the presidency. Two parties dominate: the National Liberation Party (Partido de Liberación Nacional; PLN) and the Social Christian Unity Party (Partido Unidad Social Cristiana; PUSC). The former, founded by the moderate socialist José Figueres Ferrer in 1948, was largely responsible for establishing the health, education, and welfare reforms for which Costa Rica is noted. The PUSC, a four-party coalition formed in 1977, is more conservative and business-oriented than the PLN.

System of justice

In the Costa Rican system of justice, cases may be decided by a single judge or by a panel of judges; the jury system is not used, but the courts are generally noted for their fairness. Capital punishment is banned, and sentences to the penitentiary must be for a stated number of years. The highest court is the Supreme Court of Justice. Costa Rica has no army, only a nonconscripted civil guard that has police duties. There also are district police.

Education. The constitution provides for free and compulsory education. The government oversees school attendance, curricula, and other educational matters. About one-fourth of the nation's budget is allocated to education, and literacy rates exceed 90 percent. School attendance is relatively high, with almost 90 percent of children aged 6 to 11 enrolled in primary schools and nearly 40 percent of students aged 12 to 16 enrolled in secondary schools. The University of Costa Rica has a well-planned, functional campus in San Pedro, a suburb of San José; the National University has a smaller campus in Heredia; and the "open" university (Universidad Estatal a Distancia) offers courses by television from offices in San José. The Autonomous University of Central America (founded in 1976) is also located in San José, as are several private institutions of higher education. The Technological Institute of Costa Rica, located near Cartago, provides engineering and other technical training.

Health and welfare. Costa Rica has greatly reduced the incidence of diseases associated with tropical climates. Malaria has been virtually eliminated except in the border areas with Nicaragua; water-borne diseases are rare, and mortality rates are low. The incidence of cancer and heart disease has risen, however.

Though the break between the wealthy and the manual worker is less distinct in Costa Rica than in other Central American nations, there remains a large number of agricultural and industrial workers who earn very small wages. The poorest areas are the province of Limón, the Cordillera de Talamanca, the northern lowlands, and isolated parts of the Pacific coast. The San José metropolitan area stands out as the area of greatest affluence.

Cultural life. Most Costa Rican diversions are cosmopolitan rather than nationalistic in nature. The people attend films with great frequency, enjoying international cinema. They listen to an extraordinary variety of music, especially from the many radio stations in the country. Residents of the Meseta Central attend the National Theatre, where the music played and the drama performed may come from any part of the world.

Pre-Columbian art

Costa Ricans take a strong interest in their pre-Columbian art, which includes large statues from the Pacific northwest, smaller examples of carved relief in stone from other districts, and some fine work done in the form of small objects of gold. Samples of all these may be seen in the national museum. Guayabo National Park, near Turrialba, features the country's only preserved pre-Columbian archaeological site. Genuine colonial architecture is rather scarce, the most famed example being a 17th-century mission in Orosí. Cartago's older buildings, destroyed by earthquakes, have in some cases been restored; new ones like them have also been built. Among the folk arts, Costa Rica is most famous for its highly decorative oxcarts.

The fine arts have seldom flourished in Costa Rican history, but they have received some impetus from government support, particularly with the creation in 1970 of the Ministry of Culture, Youth, and Sports. Painting, sculpting, and music all showed considerable development in the latter part of the 20th century. Particular pride was taken in the growth of the National Symphony Orchestra since 1971, with the ensemble playing large halls and also taking music to the countryside. Costa Ricans have been marginally active in the field of literature. Roberto Brenes Mesén and Ricardo Fernández Guardia were widely known as independent thinkers in the fields of education and history, respectively. Fabián Dobles has attracted international attention as a writer of novels on social-protest themes.

Costa Rica has developed the largest national park system of any Latin American country, relative to its territorial extent. These parks include a bewildering range of tropical ecosystems, such as tropical rain forest, cloud forest, dry forest, and elfin forest. Other parks include active volcanoes, turtle nesting sites, and coral reefs. The national parks are a major attraction for Costa Ricans, who flock to them on weekends and major holidays such as Easter Week, Independence Day (September 15), and the week between Christmas and New Year's Day. International tourists are also attracted by these parks, some of which are noted worldwide for their vegetation and wildlife.

Publishing

Numerous publishing houses operate in the country, issuing both fiction and nonfiction on a wide range of topics. The government-operated Editorial Costa Rica and the Editorial Universitaria Centroamericana are among the most prolific of the publishing houses. Both the number and variety of publications available in Costa Rican bookstores surpasses those of any other Central American country and some South American countries as well. *La Nación,* an independent but conservative daily, is the most widely read of Costa Rica's newspapers. It is balanced by *La Républica* and *La Prensa Libre,* independents that lean more toward reform ideas. There are several television stations, one of which is government-owned.

(F.D.P./G.S.E.)

For statistical data on the land and people of Costa Rica, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

In 1502 Christopher Columbus' fourth Atlantic voyage brought him to the shores of Costa Rica, where he remained for 18 days refitting his ships. Relations with the natives became friendly enough that they brought him a number of items of gold, possibly prompting Columbus to name the land "Rich Coast," although there is some dispute over the origin of the name. Other more promising regions forced Spain to neglect the area, however, and the few Spanish colonists clung to the coast for 60 years. In 1564 the Spanish crown established the Meseta Central village of Cartago as the first permanent settlement. Theoretically under the political jurisdiction of the captain general of Guatemala and the spiritual guidance of the bishop of León in Nicaragua, Costa Rica was ignored by both administrations. The absence of great mines meant the collection of few taxes from the Ticos, as Costa Ricans called themselves; consequently Spain provided little help in developing the infrastructure of the province. Compared with other colonies Costa Rica lacked a large labour force, so essential in the Spanish scheme of conquest. Costa Rican Indians were not numerous, and they resisted capture and disappeared into the forests rather than succumb to the encomienda, Spain's usual system of forced labour. Lacking products for a great overseas market, the Costa Ricans eked out a subsistence economy based on cacao and tobacco. Hence most people were small landowners with a close personal interest in local affairs. Historians often give credit to these developments for the growth of the democratic ideals that have become associated with Costa Rica. It should be noted, however, that some persons who became rich established a small finca- (estate-) based oligarchy which led the nation in the independence era.

Independence. When Mexico declared its independence from Spain in 1821, Costa Rica, with other parts of Central America, joined the short-lived Mexican Empire. In 1823 Costa Rica helped create the United Provinces of Central America but, disenchanted with the strife in the other four states of the federation, severed its ties in 1838. Already a pattern of isolationism had been formed, and in the many attempts to revive the federation the Costa Ricans have invariably shown little interest.

Costa Rican isolationism

From the 1840s a constant stream of oxcarts carried coffee from the Meseta Central to Pacific ports and ships bound for Europe. This trade brought British investment. Farmers with even small acreage could derive an adequate if simple existence, and the ground was laid for a society that demanded schools and roads from its government and found political participation necessary to achieve these goals.

Costa Rica's policy of isolationism did not completely save it from foreign troubles. In 1825 the province of Guanacaste seceded from Nicaragua and joined Costa Rica, creating an issue that was contended until the boundary treaty of 1896. Sharing the San Juan River with Nicaragua, Costa Rica also shared some of the canal and filibustering fevers that nearly destroyed its neighbour.

Material progress came to Costa Rica during the era of General Tomás Guardia, who dominated the nation from 1870 until 1882. His government curtailed liberty and increased the debt, but it also brought increases in coffee and sugar exports as well as widespread construction of schools. A new constitution, adopted in 1871, remained in effect until 1949. The emphasis on agricultural exports strained transportation, and, with mainly British funds, Costa Rica sought to link the Meseta Central with the seaports by railway. The chief promoter was an American, Minor C. Keith, who made a fortune with the opening of his rail line between Cartago and Limón. With vast land grants, Keith then entered the banana business. By the late 19th century, bananas were beginning to rival coffee as the chief source of Costa Rican foreign exchange, especially after Keith's investments were merged with others to form the United Fruit Company in 1899.

The last decades of the century were also marked by a gradual decline in Roman Catholic church activity in secular affairs. The Jesuits were expelled for a few years, cemeteries were secularized, and public education was expanded. In 1886 free public education became compulsory; normal schools, a museum, and a national library were founded. Though the government continued to support the church, the constitution of 1871 provided for religious toleration. Strengthening the tradition of democracy for which Costa Rica was to become famed throughout Latin America was the victory in 1890 of President José Joaquín Rodríguez in what is considered the first entirely free and honest election in all Central America.

20th century. When Nicaragua, in the Bryan-Chamorro Treaty of 1916, gave the United States permission to use the San Juan River (the border between Nicaragua and Costa Rica) as part of an interoceanic canal route, Costa Rica protested that its rights were being ignored. The claim was brought before the Central American Court of Justice, which ruled that Nicaragua had violated Costa Rican claims to the river; Nicaragua refused to accept the decision and withdrew from the court, a major factor in the court's death a year later.

Boundary dispute with Panama

Costa Rica's boundary with Panama (originally with Colombia) was also in dispute. Arbitration awards by France and the United States in 1900 and 1914, respectively, had been generally favourable to Costa Rica and were rejected by Panama. In 1921 Costa Rica attempted forcible occupation of this area (on the Pacific coast) but was diverted by the intervention of the United States. Panama then evacuated the region, but relations between the two small states were not reestablished until 1928. In 1941 the governments finally reached an accord over the boundary.

When Costa Rica held an election under direct suffrage for the first time in 1913, no candidate won a majority, and the Legislative Assembly chose Alfredo González Flores as president. General Federico Tinoco Granados, disgruntled over reforms proposed by González, in 1917 led one of the nation's few revolutions. Tinoco's despotic behaviour soon cost him his popularity. His administration was also impeded by the refusal of the U.S. government to recognize his regime, and revolts and the threat of U.S. intervention caused him to resign in 1919.

This experiment in dictatorship was not repeated, and

Costa Rica continued its tradition of democratic elections and orderly government. A literacy test for voters was adopted in 1920 and the secret ballot in 1925. On Dec. 8, 1941, the day after Japan attacked Pearl Harbor, Costa Rica declared war on Japan and on December 11 extended the declaration to Germany and Italy.

Costa Rica's most serious political crisis since 1917 came in 1948. A faction containing some alleged communists tried to prevent the seating of the president-elect, Otilio Ulate. José Figueres, a socialist landowner, put down the rebellion and turned the government over to Ulate. A new constitution, promulgated in 1949, prohibited the establishment or maintenance of an army, and the army was replaced by a Civil Guard. Elected in his own right in 1953, Figueres nationalized the banks and threatened the holdings of the United Fruit Company and several utility corporations. In 1955 he repelled an invasion by exiles residing in Nicaragua. Figueres was elected again in 1970, having meanwhile established his PLN as the dominant group in the Legislative Assembly.

In 1974 Daniel Oduber succeeded Figueres as president. Although both belonged to the PLN, they soon fell out over Figueres' ties to the U.S. financier Robert Vesco, who had found refuge in Costa Rica from an indictment on conspiracy charges in New York City. Vesco left Costa Rica in 1978, but the splintering of the PLN made possible the presidential victory of Rodrigo Carazo Odio in that year. Carazo faced serious diplomatic and economic problems. When rebellion broke out against the Somoza regime in Nicaragua, thousands fled to Costa Rica. After the overthrow of that regime in 1979, Costa Rica was further strained by the arrival of refugees escaping from civil war or civil-rights abuses in the other states of Central America. Many refugees were deported for using Costa Rica as a military base.

Economic problems

Even more enduring were economic troubles. Inflation rates fluctuated sharply, and unemployment rose. Hospital, dock, banana, and railroad workers received small raises after disruptive strikes. Economic growth slowed to near zero when the price of oil became so high that almost the entire coffee crop was needed to pay for that single import. Many years of easy credit, excessive government spending, and unfavourable trade balances brought the nation to the brink of economic ruin. Carazo and the bankers failed to reach agreement and left the problem for the new president, Luis Alberto Monge Álvarez of the PLN, who took office in 1982. In return for extending Costa Rica's debts, the International Monetary Fund and the World Bank insisted that Monge impose severe austerity measures, including devaluation of the colón, budget and tax cuts, and suspension of some subsidies.

In 1986 Monge was succeeded by another member of the PLN, Oscar Arias Sánchez, who faced many of the same economic problems. The nation continued to be beset by nearly $5 billion in foreign debt, too-rapid urbanization, inadequate housing, unemployment, and adjustments necessitated by privatization of state monopolies. More than one-third of the nation's income was derived from international loans. Civil wars elsewhere in Central America continued bringing thousands of fugitives into Costa Rican exile, and illicit drug traffic imposed new forms of corruption on the land.

Aside from some easing of the debt structure, Arias could accomplish little with these matters, but he created for himself a powerful role in international affairs. A wealthy coffee grower and political scientist, Arias spent most of his term leading a regional peace movement designed to end the bitter strife in El Salvador, Nicaragua, and Guatemala. Although often receiving little support from the United States, he achieved considerable success and did more than any other person to reduce tensions and create machinery to end the bloody Central American struggles. In 1987 President Arias received the Nobel Prize for Peace for his efforts. Arias, constitutionally ineligible to run in 1990, was succeeded by Rafael Angel Calderón Fournier of the opposition PUSC. (T.L.K.)

For later developments in the history of Costa Rica, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

El Salvador

The Republic of El Salvador (Spanish: República de El Salvador), with a total area of 8,124 square miles (21,041 square kilometres), is the smallest and most densely populated of the Central American countries. Its territory is situated entirely on the western side of the Central American isthmus, and it is therefore the only Central American nation that lacks a Caribbean coast. El Salvador is bounded to the south by the Pacific Ocean, to the northwest by Guatemala, and to the north and east by Honduras. The capital is San Salvador. Despite having little level land, El Salvador is largely an agricultural country, heavily dependent upon coffee exports. Since the late 1970s the nation has been a focus of international attention owing to its continuing civil war and external involvement in the country's intense internal conflicts.

PHYSICAL AND HUMAN GEOGRAPHY

The volcanic axis

The land. *Relief and soils.* The entire territory of El Salvador is situated on the Central American volcanic axis, which determines the major geographic regions of the country. Relief is dominated by the central highlands, consisting largely of a west-east line of volcanoes, many still active, that crosses the centre of the country. This volcanic range includes 20 cones, from the westernmost Izalco Volcano (6,266 feet; 1,910 metres), through those of San Salvador (6,430 feet) and San Miguel (6,988 feet), to that of Conchagua (4,078 feet) in the extreme east. Along this line flooded volcanic craters provide the country's largest lakes at Coatepeque (15 square miles; 39 square kilometres), Ilopango (25 square miles), and Olomega (20 square miles). These volcanoes are separated by a series of basins (commonly referred to as El Salvador's central plain), situated at elevations of between 3,500 and 5,000 feet (1,000 and 1,500 metres), whose fertile soils, derived from volcanic ash, lava, and alluvium, have for centuries supported large concentrations of population.

To the south, where these central highlands fall away to the Pacific coast, is a narrow coastal plain with average elevations of between 100 and 500 feet (30 and 50 metres). Its level, fertile soils, deposited by the numerous small rivers draining from the central highlands, combined with high year-round temperatures and abundant rainfall, provide favourable conditions for plant growth and agriculture. North of the central highlands, and parallel to them, a broad interior plain drained by the Lempa River is situated at elevations between 1,300 and 2,000 feet (400 and 600 metres). Intermittently broken by ancient dormant volcanic structures, and adversely affected by poor drainage and high soil acidities, this interior plain has provided a less attractive environment for human habitation.

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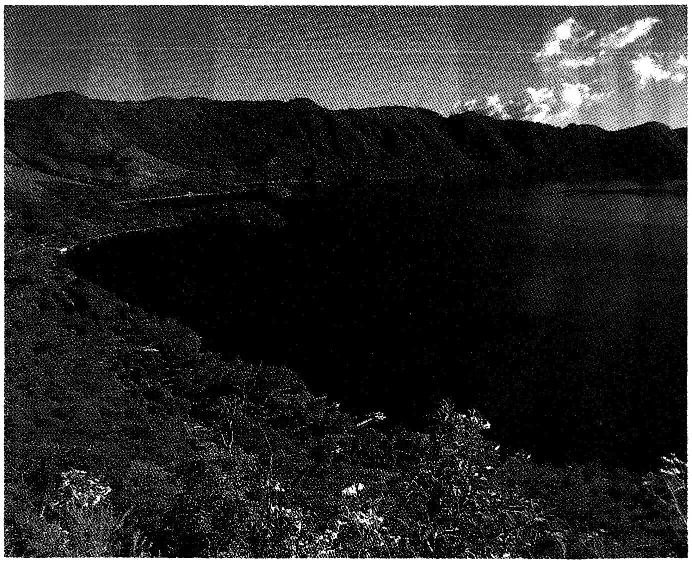

Lake Coatepeque, one of the largest of a line of flooded volcanic craters in western El Salvador.

Extending along the entire northern border region are a range of highlands, with average elevations of 5,000 to 6,000 feet, formed by ancient and heavily eroded volcanic structures. The steepness of slope, excessive forest clearance, and overuse of soils have led to serious deterioration of the environment of this northern region. In the extreme northwestern part of the country there are limited outcrops of limestone rock associated with the older nonvolcanic structures of Honduras.

The Lempa river system

Drainage. Two principal river systems and their associated tributaries drain the major part of the country. Most important is the Lempa, which enters El Salvador from Guatemala in the northwestern corner of the country and flows eastward for 80 miles (130 kilometres) across the interior plain to form part of the border with Honduras before turning sharply south to run 65 miles (105 kilometres) through the central highlands and across the coastal plain to its mouth on the Pacific. The Lempa was navigable for several miles inland prior to the construction of two major hydroelectric installations on its middle reaches: the Cerron Grande and First of November dams. The eastern part of the country is drained by the Rio Grande de San Miguel system. A series of short north-south streams drain directly from the central highlands to the Pacific.

Climate. The climate of El Salvador is tropical but is moderated by altitude in the interior; in general it is warm rather than hot, varying between 59° and 73° F (15° and 23° C). Heavy rains, known as the *temporales,* fall in the winter season from May to October. The dry summer season lasts from November to April. There is considerable climatic variation in the different regions. The Pacific lowlands and low areas in the middle Lempa River valley have mean monthly temperatures between 78° and 85° F (25° and 29° C). At San Salvador, the capital, which is 2,156 feet above sea level, the maximum monthly mean temperature is 94° F (34° C) in March, and the lowest monthly mean is 60° F (16° C) in January. In the mountains, above 4,800 feet, mean monthly temperatures vary between 63° and 72° F (17° and 22° C). Annual precipitation on the Pacific lowlands averages about 68 inches (1,700 millimetres); on the southern and northern mountain ranges, at heights between 2,000 and 3,500 feet, the average is between 70 and 97 inches (1,800 and 2,500 millimetres); the higher mountains receive a little more. Annual precipitation recorded in the deeper valleys and surrounding plateaulike areas is between 45 and 60 inches (1,100 and 1,500 millimetres).

Plant and animal life. The higher mountain regions have temperate grasslands and the remnants of deciduous oak and pine forests. On the central plains and in the valleys small deciduous trees, bushes, and subtropical grasslands are found. The coastal plains and the lower slopes of the southern mountains are covered with either savanna (parklike grassland) or deciduous forests. Among the many species of trees is the balsa, known for its beauty and soft perfume; its trunk yields excellent lumber as well as resin that is used in antiseptics and medicinal gums.

Trees and tree products

Valuable wood also is obtained from the cedar, mahogany, laurel, nispero, and madrecacao trees and is used for the manufacture of furniture. Particularly beautiful is the maquilishuat, the pink-tufted national tree of El Salvador. Several species of palm and coconut trees grow in the coastal zone, and there are many varieties of tropical fruit, such as coconut, tamarind, melon, watermelon, and mango. The izote is the national flower.

Because of the amount of land under cultivation, El Salvador is considerably less rich in animal life than most Central American countries. Rodents, reptiles, and insects of many kinds, however, are common. There is a wide variety of birdlife, which includes wild duck, the white and the royal heron, the urraca (which has a blue breast and a grey head and is known for its call resembling a scoffing laugh), the blue jay, and many more, some of which have fine plumage. A wide variety of fish, as well as turtles and crocodiles, inhabit the streams, lakes, and rivers.

Settlement patterns. More than three-fourths of the present population live in the intermontane basins of the central highlands. For millennia before the Spanish conquest these areas supported large agricultural communities of Amerindians dependent on the cultivation of indigenous crops, such as corn (maize), beans, and squashes. The ruins of their cities are evident at sites such as Chalchuapa, Sihuatán, and Cara Sucia. The major Spanish colonial settlements, which became the nation's principal cities, were also situated in these central basins: Santa Ana, Ahuachapán, San Salvador, San Vicente, and San Miguel. This concentration of population was perpetuated during the colonial period by the commercial production of indigo and sugar on private estates, owned by a few wealthy families, alongside the continuing indigenous subsistence farming of Amerindian peasants living in poverty. From the 19th century these basins and their surrounding slopes provided an ideal environment for the cultivation of coffee, which became the basis of the national economy. In the 20th century urban growth and industrialization increased the concentration in the highland centre of the country. More than one-half of the national population now live in the central cities. This distribution of population exacerbates the effects of natural disaster in a zone prone to seismic activity, and most of these cities have been subject to one or more destructive earthquakes.

San Salvador, the capital, was founded in 1525 and in the 20th century has grown rapidly to absorb many surrounding settlements. This major conurbation now stretches continuously from Nueva San Salvador in the west to Lake Ilopango in the east. It contains an estimated one-fifth of the total population. In the east San Miguel, located on the slope of the volcano San Miguel, is a thriving city where the old Spanish colonial and modern life-styles merge. The city of Santa Ana is the commercial centre of western El Salvador.

Central highlands emigration

The overpopulation in the central highlands has resulted in out-migration to the coastal plain, which, since 1945, has been transformed by extensive cotton and cattle farming. Past settlement in the northern highlands has caused severe deforestation and soil degradation in a region of steep slopes and poor soils. The region remains overpopulated by dispersed peoples, the majority of whom are restricted to subsistence farming.

The people. Before the Spanish arrived in Salvadoran territory in the 16th century, it was occupied by a complex of Indian tribes. Of these the Pocomam, Chortí, and Lenca, all related to the Maya, were the more ancient, but the Pipil, whose civilization resembled that of the Aztecs in Mexico, were predominant. The Pipil name for their territory as well as for their capital was Cuscatlán, meaning "Land of the Jewel"; the name is still sometimes applied to El Salvador today. Archaeological ruins dating from Indian times are to be seen at Tazumal, Pampe, El Trapito, and San Andrés. Of several large towns founded by the Indians, Sonsonate and Ahuachapán still exist.

The intermarriage of Spanish settlers with the local population has resulted in a racially homogenous people. Almost nine-tenths of the contemporary population is mestizo; the remainder is Indian (notably the Izalco Indians and the Pancho from the village of Panchimalco near the capital) and white.

During the pre-Spanish epoch various Indian dialects were spoken, the most important of these being Nahuatl, spoken in the central region of the country, and Poton, spoken in the east. After the initial conquest, Spanish became the official language, and the Indian dialects slowly fell into disuse. A government effort is being made to preserve Nahuatl. Most Salvadorans profess the Roman Catholic religion.

Severe economic conditions complicated by the civil war beginning in 1979 caused dramatic changes in El Salvador's demographics. It is estimated that about 20 percent of the population left the country, departing in about equal numbers for neighbouring countries and the United States. Within the remaining population there has been massive displacement characterized by a general movement of people from the conflict zones in the north and east to the central cities. The disruption of lifeways and the emigration of many young Salvadorans has brought an accompanying decline in the rate of natural increase. Nevertheless, overcrowding is a severe problem.

The economy. The economy is predominantly agricul-

Primacy of agriculture

tural, although there was rapid industrial expansion in the 1960s and '70s. Despite its concentration on agriculture the country has trouble feeding itself and must import food. The main cause is the disproportionate distribution of land in favour of commercial crops, leaving many of the peasants landless and unable to grow subsistence crops.

The most important agricultural products are coffee, cotton, corn, and sugarcane. Coffee alone accounts for a substantial part of the value of the total agricultural production and more than half of the value of exports; the biggest buyers are the United States, Germany, The Netherlands, and Japan. Cattle raising is also an important activity. While some woods are used for furniture making, as mentioned above, much of it is also used for fuel; other forest products are dyewoods and balsam, a medicinal gum of which El Salvador is the world's main source.

Commercial fishing, regulated by the government, has added to the country's export earnings. Most of the fish caught commercially or for sport come from offshore waters and coastal lagoons; they consist chiefly of crustaceans (including lobster and shrimp), mullet, snappers, jacks, groupers, sharks, and anchovies. There is no mineral exploitation of significance. The main power sources, meeting most of the country's needs, are the hydroelectric projects on the Lempa River 35 miles northeast of San Salvador, which are administered by a government agency.

In the middle and late 20th century there was a steadily increasing investment in industry, stimulated by the Central American Common Market. Industrial plants rose throughout the country, and existing facilities were expanded, helped by government incentives, an advanced banking system, and development credits from abroad. In 1979, however, manufacturing began a serious decline, which continued into the 1980s. Civil unrest and political instability were the primary causes. Manufactures include beverages, canned foods, organic fertilizers, cement, plastics, cigarettes, shoes, cotton textiles, leather goods, petroleum products, and machinery.

Tourism

Tourism was an increasing source of income but also suffered a decline after 1979. In 1980 the nation's commercial banks and the export marketing agencies were nationalized. Government revenue is mostly raised by sales taxes and by customs and export duties on coffee. Since 1980 the government has relied heavily on large aid disbursements from the United States.

El Salvador has adequate transportation facilities except in some of the remoter areas. Two main routes of the Pan-American Highway system cross El Salvador from Guatemala to Honduras, forming the framework of a road system that reaches almost all parts of the country; one of these routes runs across the central highlands, the other across the coastal plains. Several paved all-weather roads connect with these main highways. The country's narrow-gauge railroad is operated by a national agency; the main tracks link the capital with ports on the coast and with the Guatemalan border. For sea commerce, El Salvador relies on three ports—Acajutla, La Libertad, and Cutuco (near La Unión). El Salvador's outlet to the Atlantic is through the Guatemalan port of Puerto Barrios, with which San Salvador is linked by road and rail, via Guatemala City. A new international airport was constructed in the 1970s on the coastal plain 25 miles south of the capital. It replaced Ilopango airport, which now serves as a military base. Severe damage to the nation's transportation network resulted from the civil war.

Administration and social conditions. *Government.* The constitution of 1983 provided for a representative form of government with three branches—legislative, executive, and judicial. All Salvadoran men and women who have attained the age of 18 have the right and duty to vote. Executive power is exercised by the president, who serves a nonrenewable five-year term, the cabinet ministers, and the undersecretaries of state. Legislative power is vested in a unicameral, popularly elected Legislative Assembly, which also has the power to appoint the president and vice presidents, if no candidate has an absolute majority in an election, and to veto presidential appointments; its members serve three-year terms. The judicial branch is composed of a Supreme Court of Justice, whose magistrates are selected by the Legislative Assembly, and of other tribunals as established by statute.

Local administration

El Salvador's territory is divided into 14 *departamentos,* each of which is divided into *distritos* (districts), which are further divided into *municipios* (municipalities). Each department has a governor and a substitute governor, appointed by executive power; and each municipality has a popularly elected municipal council composed of a mayor, a secretary, and aldermen, the number of whom are in proportion to the population.

Education. All public and private institutions of learning are under the jurisdiction of the Ministry of Education. Since 1968 the school system has been composed of preschool, primary, and secondary educational categories, followed by university-level education. Primary education is free and compulsory, but the illiteracy rate is relatively high. Among the institutions of higher education are the University of El Salvador, the University Dr. José Matías Delgado, and the Central American University José Simeón Cañas. There are also schools for technology, fine arts, agriculture, social services, and nursing.

Social welfare and health. The traditionally low standard of living of artisans and agricultural labourers has worsened as a result of the civil war. More than 60 percent of the economically active population are unemployed or underemployed. Attempts are being made by the government to achieve a more equitable distribution of income through a major program of agrarian reform and through education, vocational training, and community projects. Progress, however, has been exceedingly slow, particularly in the area of land reform, which has been resisted by the powerful landowning elite. Low-cost housing, medical assistance, and employment programs have been introduced in an attempt to meet the needs and problems of the displaced and the unemployed. Such programs, however, have difficulty keeping up with deteriorating conditions. The doctor-to-patient ratio is low and falling, and most doctors serve urban areas. The war and population displacement have caused the reappearance and spread of diseases, particularly malaria, in many areas; malnutrition is increasingly prevalent.

Cultural homogeneity

Cultural life. The cultural life of El Salvador, like its population, is an amalgam of Indian and Spanish influences. There is greater cultural homogeneity than in any other Central American country, with Spanish spoken universally and Indian customs (though little spoken language) surviving in small clusters of villages, such as those around Izalco and Nahuizalco, and traditional crafts being maintained in Ilobasco (pottery) and Izalco (textiles). The present indication of this cultural mix is represented in the country's rich tradition of folklore, poetry, and painting.

The Roman Catholic church has been a major influence on almost every aspect of cultural life. In higher education, for example, the earliest institution, the Colegio de la Asunción, was founded in 1841 to teach Latin and theology, while the nation's leading contemporary university, Central American University, is a Jesuit foundation.

Private benefactors have played as important a role as that of the government in patronizing the arts. San Salvador's principal theatre was built by public subscription; the majority of newspapers and publishers are privately owned; the major national parks were established by private funds; and the principal museum originally was a private foundation. More recently the government has increased its contribution to national cultural life, particularly in its expansion of secondary and advanced education. (R.S.V./D.G.B.)

For statistical data on the land and people of El Salvador, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

The colonial period. The Spanish conquest and colonization of El Salvador began in 1524 with the arrival of an expedition from Guatemala led by Pedro de Alvarado. Alvarado's troops met determined opposition from a Nahua tribe, the Pipil, who occupied much of the region west of the Lempa River. However, superior tactics and armaments enabled the Spaniards to push on to the Pipil capital

The Spanish conquest

of Cuscatlán. Alvarado soon returned to Guatemala, but a second expedition in 1525 founded a Spanish town called San Salvador near the site of Cuscatlán. Pipil warriors forced the Spanish settlers to withdraw, however, and the community would be resettled several times before it was permanently established in 1528.

Thereafter, the town of San Salvador would serve as the capital of a province of the same name that included most of the eastern three-fourths of the territory of modern El Salvador. The area to the west (comprising the present-day departments of Sonsonate, Santa Ana, and Ahuachapán), which the Pipil called Izalcos, was organized in 1558 as the autonomous province of Sonsonate and would not be incorporated as a part of El Salvador until 1823.

The lands that would form El Salvador became the agricultural heartland of the captaincy general of Guatemala. Although most of the inhabitants were obliged to depend on subsistence farming, the more fortunate Spaniards found wealth in the export of a variety of local products, all of which experienced periods of "boom and bust." Cocoa was the most important source of wealth during the 16th century. Increased competition from other colonies led to a marked drop in revenue from cocoa by 1590, and the following century was clearly a period of stagnation for the region. Recovery in the 1700s came as a result of increased exports of indigo.

The indigo trade led to the development of a fairly sophisticated form of commercial agriculture and the creation of large estates operated by families whose members played a leading role in provincial affairs. For the native people, however, the indigo boom chiefly meant that an additional burden was placed on an already exhausted work force.

Indigo planters' independence role

Independence. A variety of considerations caused the Salvadoran indigo planters to take a leading role in agitating for Central American independence. These included the hard times caused by a sharp decline in indigo production during the first decade of the 19th century, a long-held hostility toward Guatemalan merchants who controlled much of the economy of San Salvador, and the conviction that the province should be organized as a bishopric so that it need no longer depend upon the archbishop of Guatemala for pastoral services.

In November 1811 the arrest of a member of one of the planter families ignited an uprising led by José Matías Delgado, the provincial vicar of San Salvador, and his nephew Manuel José Arce. The rebels held the government for nearly a month before Spain's authority was restored by the captain general of Guatemala, whose measures seemed more conciliatory than repressive. A second, shorter uprising in 1814 had wider popular support, and it provoked a more severe response from the captain general, costing Arce more than four years in prison.

Conflict with Mexico

In 1821 the province endorsed Guatemala's declaration of independence from Spain. The Salvadorans, however, opposed the Guatemalan decision to accept incorporation into Agustín de Iturbide's Mexican Empire, a stance that led to confrontations with Guatemalan and Mexican armies. Faced with defeat late in 1822, a Salvadoran congress sought adoption of a resolution providing for the province's annexation to the United States; but this scheme was abandoned when Iturbide's government collapsed in 1823. Meeting in June of that year in Guatemala City, a Central American constitutional convention chose Delgado as its president, appointed Arce as a member of the provisional executive triumvirate, and went on to draft a constitution, which was completed in 1824. The state thus created was now called the Federal Republic of Central America, having earlier been termed the United Provinces of Central America; in 1825 Arce became its first president.

Nationhood. The state of San Salvador (the modern name, El Salvador, was not used until 1841) played an important part in the affairs of the Central American federation. Not only was it the birthplace of the federation's first president, it was also there that a revolt was sparked against Arce in 1827, beginning the civil war in which Central American Liberals and Conservatives contested for control of the new nation. This conflict, which caused the collapse of Arce's presidency, ended in 1829 with the seizure of the federal government by Francisco Morazán, commander of the Liberal army. Having cast their lot with Morazán, the Salvadorans became his most loyal allies and were rewarded in 1834 with the transfer of the federal capital to the city of San Salvador. The Salvadorans were so attached to the ideals of federation that the state did not assume sovereign powers until 1841, after the other four states had already left the union.

Sovereignty did not signal the arrival of peace and prosperity for El Salvador; if anything, the nation experienced increased civil strife and international conflict for several decades after 1841. From that year until 1863, just one chief of state could claim continuous service that ran two full years. During this time, El Salvador was involved in wars with neighbouring countries that usually arose from attempts to meddle in their politics. Often El Salvador found that the final arbiter of its political affairs was Rafael Carrera, Conservative dictator of Guatemala from 1839 until his death in 1865. In the midst of this turmoil, the nation secured the establishment of the long-sought bishopric and saw the beginnings of the coffee industry, which was advanced in part by the policies of President Gerardo Barrios Espinosa (1861–63).

Liberal reforms

A coffee republic. The presidency of Francisco Dueñas (1863–71) pointed toward greater political stability in the nation; real change came, however, when his overthrow in 1871 marked the beginning of 60-year period of rule by Liberals who focused on the pursuit of economic growth and domestic tranquility. Late in the 19th century, a substantial shift in the nation's economy became essential when the development of synthetic dyes severely reduced the income normally generated by the export of indigo. Salvadorans solved this problem by means of a "coffee revolution." New lands had to be opened to cultivation, a step facilitated during the administration of Rafael Zaldívar (1876–85), who authorized the sale of Indian lands. These proceedings provoked Indian uprisings, which were put down by a newly created rural mounted police force.

The coffee planters developed a highly efficient system of plantation enterprises and formed a closely knit elite that used its growing economic strength to ensure that the government served its interests. Among the small number of controlling families, just two—the Meléndez and Quiñónez families—monopolized the office of president between 1913 and 1927.

Military dictatorships. The coffee barons' direct control of the presidency ultimately came to an end as a consequence of the worldwide depression that began in 1929. A coup installed General Maximiliano Hernández Martínez as president in December 1931 and initiated a succession of military governments that controlled the nation through 1979.

La matanza

The persistence of military rule can be partly explained as a result of a two-day revolt by farm workers in January 1932 that was organized by Augustín Farabundo Martí, head of the recently formed Salvadoran Communist Party. Hernández Martínez easily suppressed the rebellion and authorized the summary execution of at least 10,000 suspected participants. The uprising and its brutal repression, which is referred to as *la matanza* ("the slaughter"), were momentous events in the history of the nation. The revolt demonstrated the value of the military dictatorship to the landed elite, which became convinced of the need for eternal vigilance against the menace of a communist revolution. It also eliminated the immediate threat from the left as well as most of the last vestiges of Indian culture.

Personally honest and austere, Hernández Martínez sought to emulate the fascist dictators of Europe, but he may be best known for his interest in occult arts. His regime survived a coup in April 1944, but the following month a general strike launched by university students brought the nation to a standstill and caused the dictator to resign from office. There was no real change, however, until 1948, when a revolt by young army officers installed a junta headed by Major Oscar Osorio. This "Majors' Revolution" gave rise to policies and patterns of behaviour that would have a central role in the practice of Salvadoran politics during the next 30 years.

Elected to a six-year term as president in 1950, Osorio organized the Revolutionary Party of Democratic Unification (Partido Revolucionario de Unificación Democrática; PRUD) and launched a variety of reform projects, such as the development of hydroelectric facilities and urban housing projects. He also extended collective bargaining rights to urban workers; but, for the most part, the reforms served to encourage economic growth and to benefit the middle class. Osorio's successor, Lieutenant Colonel José María Lemus (1956–60), continued these programs, but there was no improvement in the living standards of workers. When faced with open discontent, Lemus resorted to repressive measures, and a military coup deposed him in October 1960.

A second coup in January 1961 brought Lieutenant Colonel Julio Adalberto Rivera (1962–67) to power. PRUD was dismantled and replaced by the National Conciliation Party (Partido de Conciliación Nacional; PCN), which would control the national government for the next 18 years. Under the banner of the Alliance for Progress, Rivera advanced programs aimed at economic growth and diversification which enabled El Salvador to take advantage of the increased trade opportunities offered by the recently formed Central American Common Market (CACM). A greater degree of political liberty seemed evident from the rise of the Christian Democratic Party (Partido Demócrata Cristiano; PDC) and the victory of its candidate, José Napoleón Duarte, in the 1964 mayoral election in the city of San Salvador. At the same time, the Rivera government oversaw the formation of the Democratic Nationalist Organization (Organización Democrática Nacionalista; ORDEN), a large, secretive paramilitary organization.

The Soccer War

Colonel Fidel Sánchez Hernández (1967–72) encountered difficulties as a result of the decline in world prices for coffee and cotton, but in 1969 the nation's attention was diverted from economic problems by the outbreak of the Soccer War (Guerra de Fútbol) with Honduras. This conflict broke out shortly after the two nations had played three bitterly contested matches in the World Cup competition, but the real causes for the war lay elsewhere.

In the first place, there was a long-standing dispute concerning the location of portions of the border between the two countries. Also, Hondurans resented the substantial trade advantage El Salvador held over them on the basis of the rules of the CACM. Most important was the problem raised by some 300,000 Salvadorans who had migrated to Honduras in search of land or jobs and who now found themselves threatened by an involuntary repatriation program begun by the Honduran government. Spurred by reports of the mistreatment of these refugees, the Salvadoran government opened hostilities on July 14, 1969. A ceasefire took effect on July 18, but El Salvador continued to force the action until the OAS threatened economic sanctions against the country on July 29. The brief war had cost several thousand lives, and a peace treaty between the two nations was not concluded until 1980.

Continuing economic troubles and the growing popularity of José Napoleón Duarte, the PDC candidate who headed a coalition slate, suggested that the military might lose control of the government in the 1972 elections. Members of ORDEN supervised voting in the outlying provinces, however, and managed to assure the victory of the PCN's Colonel Arturo Armando Molina. An attempted coup afterward achieved little more than Duarte's arrest and exile to Venezuela, where he resided until 1979.

During the period that encompassed the tenure of President Molina (1972–77) and that of his successor, General Carlos Humberto Romero (1977–79), the nation experienced more frequent expressions of public discontent and growing abuses of human rights. The growing opposition among Roman Catholic clergymen to the church's traditional defense of the status quo provided one clear sign of widening concern about the problem of social injustice in El Salvador. This period also witnessed the formation of mass popular front organizations that expressed the grievances of peasant groups and labour unions by such means as strikes, demonstrations, and parades. By 1979 several separate guerrilla organizations were operating in El Salvador.

Increasing government repression

Apart from an agrarian reform proposal that was offered (and quickly retracted) in 1976, the government had no response to this opposition other than tightening the screws of repression. Peasants surely suffered the brunt of efforts to stifle dissent, but the most egregious example of government violence came in 1975, when at least a dozen university students were shot to death while protesting the use of public funds to hold the Miss Universe contest in El Salvador. The political situation steadily worsened until Romero was removed from office by a military coup in October 1979.

Civil war. By the time of Romero's ouster, the nation was plunged into a civil war that would drag on into the 1990s. There were other significant consequences to be noted. Most obvious was the military's loss of the monopoly it had held on the direct exercise of governmental authority for nearly 50 years. At the same time, there was a change in the relationship between the military and the nation's propertied elite. The latter group felt it could no longer rely entirely on the armed forces for protection and sought to broaden its base of support by the formation in 1981 of a new political organization, the Nationalist Republican Alliance (Alianza Republicana Nacionalista; ARENA) led by Roberto D'Aubuisson Arrieta.

In addition, the role of the United States, which previously had shown very little interest in the affairs of El Salvador, changed markedly with Ronald Reagan's inauguration as president in January 1981. During the balance of the decade, the United States supplied El Salvador with financial aid amounting to $4 billion, assumed responsibility for the organization and training of elite military units, supported the war effort through the provision of sophisticated weaponry, and used its influence in a variety of ways to guide the political fortunes of the nation.

The years following Romero's downfall provided a kaleidoscopic array of events. The governing junta formed in October 1979 collapsed three months later when its civilian members resigned because of their failure to reach agreement on reforms and their inability to bring the military under control. José Napoleón Duarte returned from exile and became head of the second junta, which enacted a package of laws that included an agrarian reform program. Reforms did not mean that there would be any reduction in the level of political violence. That was made clear in March 1980 when Archbishop Oscar Arnulfo Romero, who had become a vehement critic of the military establishment, was assassinated while offering mass. The message was reiterated at the end of the year, when the military murdered three American nuns and a Roman Catholic lay worker.

By that time, the guerrilla units had joined in a single organization, the Farabundo Martí National Liberation Front (Frente Farabundo Martí de Liberación Nacional; FMLN), and announced the opening of a "final offensive" in January 1981. The offensive was by no means final, however, and the fortunes of the guerrilla army would ebb and flow throughout the balance of the decade. During that time, the guerrillas initiated and survived hard-fought battles with government troops who were trained and supplied by the United States.

Election of 1982

Elections held in 1982 enabled the formation of a constituent assembly that organized a provisional government and drafted a new constitution (the third since 1948) which was promulgated in December 1983. Duarte was elected president the following March. Although a meeting held with guerrilla leaders in the fall of 1984 raised hopes that Duarte could negotiate an end to the civil war, the talks led nowhere; furthermore, his presidency was plagued by misfortune. He made no progress in his efforts to achieve peace or advance social and economic reforms. At the end of his term, charges of widespread corruption in the government contributed to the victory of the ARENA candidate, Alfredo Cristiani, in the 1989 presidential election. Duarte died of stomach cancer shortly after Cristiani's inauguration.

Cristiani continued to enforce harsh strictures on dissent, but he also showed willingness to examine FMLN proposals for peace. The war endured into the 1990s, however, by which time at least 70,000 people (mostly noncombatants) had lost their lives, the economy was in shambles, and massive damage to the infrastructure was evident everywhere. Clearly, the nation was far from the goals of peace, prosperity, and social justice that it had long sought. (P.F.F.)

For later developments in the history of El Salvador, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Guatemala

The Republic of Guatemala (Spanish: República de Guatemala) lies within the western portion of the isthmus connecting North and South America. Its land borders are with Mexico to the west and north, with Belize to the northeast, and with Honduras to the east and El Salvador to the southeast. The Pacific Ocean marks its southern shores, and a short coastline in the northeast opens upon the Caribbean Sea. With an area of 42,042 square miles (108,889 square kilometres), it is one of the larger nations in Central America and the most heavily populated. Its capital, Guatemala City, is the region's major metropolitan centre. The dominance of an Amerindian culture within its interior uplands distinguishes Guatemala from its neighbours. Politically, Guatemala has a long history of unstable governments prone to frequent and often violent military overthrows.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* The surface of Guatemala is dominated by four major features. In the south is a landscape of volcanic origin that extends for a distance of 180 miles (290 kilometres) between neighbouring Mexico and El Salvador. It is bordered on its seaward margin by the Pacific Ocean. The Petén *departamento,* a large, somewhat rectangular area, juts northward to occupy a portion of the Yucatán Peninsula, a limestone platform shared with Mexico and Belize. Sandwiched between the volcanic landscape and the Petén are the high ranges and valleys of the sierras of northern Central America. These arc gently eastward from Mexico for a distance of 210 miles, extending into northern Honduras and continuing on beneath the surface of the Caribbean.

The southern volcanic region

The southern volcanic region of Guatemala consists of three elements: a row of volcanoes of geologically recent origin, flanked by a deeply eroded volcanic tableland of older origin to the north and a narrow coastal plain constructed of volcanic debris on the south. The alignment of volcanic cones begins with the Tacaná Volcano (13,428 feet [4,093 metres]), located on the frontier with Mexico, and continues eastward across Guatemala into El Salvador. The highest peak is Tajumulco (13,845 feet [4,220 metres]). Among these are three continuously active volcanoes—the growing summit of Santiaguito located on the southern flanks of Santa María (12,375 feet [3,772 metres]), Fuego (12,582 feet [3,835 metres]), and Pacaya (8,371 feet [2,552 metres]).

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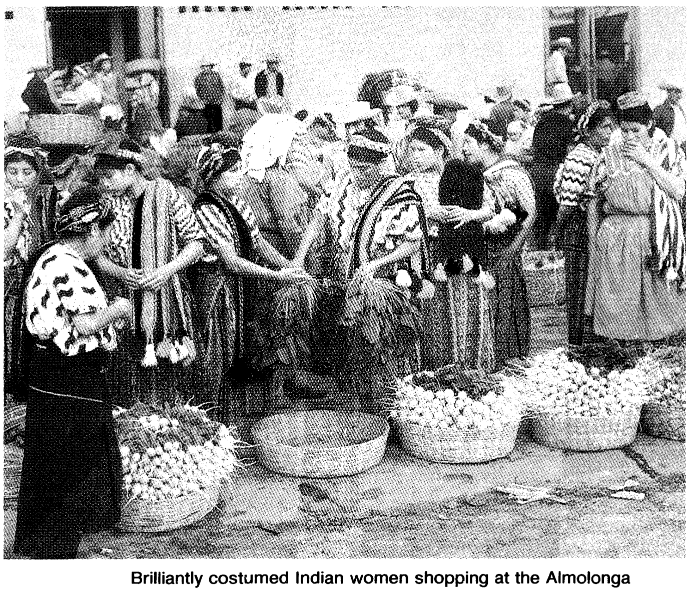

Brilliantly costumed Indian women shopping at the Almolonga market in the western highlands of Guatemala, near Quezaltenango.

The rugged and deeply dissected volcanic tableland, which lies to the north of the volcanic row, averages 9,000 feet in elevation near the Mexican border and declines gradually to 3,000 feet at the opposite border with El Salvador. Ash-filled basins and a number of scenic lakes are scattered throughout this region. From the base of the volcanic row, the Pacific coastal plain slopes south to the shoreline of the Pacific Ocean. Rising imperceptibly from sea level to an elevation of 1,500 feet, the plain extends east-west for a distance of about 150 miles. Ranging between 25 and 50 miles in width, it is one of the nation's richest agricultural areas. Three-fourths of the population and most of the major cities are concentrated there, and the volcanic eruptions and earthquakes characteristic of this region have repeatedly taken a heavy toll of property and life.

The range and valley landscape of the sierras provides a major barrier between the heavily occupied volcanic landscape to the south and the sparsely populated Petén to the north. Sierra los Cuchumatanes to the west rises to elevations in excess of 10,000 feet. Eastward, the lower sierras of Chamá, Santa Cruz, Chuacús, Las Minas, and the Montañas del Mico are separated by deep valleys that open eastward upon a narrow Caribbean shoreline.

The Petén, lying largely below 1,000 feet in elevation, exhibits a knobby, or hilly, surface dominated by subsurface drainage of water. The region is replete with scattered lakes, Lake Petén Itzá being the largest. Extensive flooding takes place during the rainy season.

Drainage. The east-flowing Motagua River and west-flowing Cuilco pass in opposite directions through a structural trough that serves as the boundary between the volcanic terrain of southern Guatemala and the sierras that extend across the midsection of the nation. South of this trough, the land is drained by a series of small rivers that drain into the Pacific Ocean.

The sierra region is drained by larger rivers that flow primarily north into the Gulf of Mexico by way of the Usumacinta River. The 250-mile-long Motagua River is the longest of a series of rivers draining eastward toward the Caribbean. Because of the nature of the bedrock in the Petén, much of the region is drained by the subsurface flow of water.

Erosion problems

Soils. The volcanic belt of southern Guatemala contains some of the nation's most productive soils; nevertheless, the northernmost sector of this region is particularly subject to erosion induced by the prevalence of steep slopes, and deforestation. Within the sierra region, heavier rainfall combined with thinner soils on the steep slopes and the wanton destruction of forests have led to widespread erosion. The limestone surface of the Petén provides for shallow and stony soils that have greatly restricted the development of agriculture.

Climate. Located within the tropics and with elevations ranging between sea level and more than 13,000 feet, Guatemala experiences a diversity of climates. Below 3,000 feet in elevation average monthly temperatures range between 70° and 80° F (21° and 27° C) throughout the year; between 3,000 and 5,000 feet (900 and 1,500 metres) temperatures range between 60° and 70° F (16° and 21° C); and from 6,000 to 9,000 feet (1,500 to 2,700 metres) between 50° and 60° F (10° and 16° C). Above 9,000 feet temperatures are marginal for crops, but the grazing of animals is possible.

Near-desert conditions prevail in the middle section of the Motagua River valley, whereas precipitation in excess of 150 inches (3,800 millimetres) occurs at higher elevations of the Pacific-facing volcanic row and on the north- and east-facing slopes of the sierras. In general, a dry season prevails between November and April; however, moisture-laden trade winds from the Caribbean yield rainfall throughout the year on north- and east-facing slopes. In general, an average of 40 to 80 inches (1,000 to 2,000 millimetres) of precipitation is received in southern and eastern Guatemala, but this is doubled in areas located nearer the Caribbean shoreline.

Severe tropical storms, especially during the months of September and October, often deluge the country with damaging floods. Strong winds accompanying these storms as well as winter invasions of cold air occasionally place crops at risk.

Plant and animal life. In the Petén, a dense rain forest is interspersed with patches of savanna grasslands. The sierras are forested with oak and pine. Except for the highest slopes in the volcanic highlands, stands of pine, fir, and oak have been largely destroyed. On the Pacific coastal plain, the landscape largely has been cleared of its tropical forest and savanna.

Lowland fauna

The richest variety of animal life inhabits the lowland forest areas, although some species, such as deer, monkeys, peccaries, tapirs, ocelots, and jaguars, are increasingly rare. The birdlife of the rain forests is particularly exuberant and includes the radiantly plumaged quetzal, the national bird, for which a reserve has been set aside in the sierras near Cobán. Among the reptiles of note are numerous snake species, crocodiles, and iguanas.

Settlement patterns. Approximately 60 percent of the population of Guatemala is concentrated within the volcanic uplands and adjacent Pacific coastal plain to the south and west of Guatemala City. A little more than 10 percent live to the east and south, while only about 2 percent reside within the Petén. The remaining populace resides in the region of the sierras. Of the urban dwellers who make up some 40 percent of the population, nearly one-half inhabit the metropolitan area of Guatemala City.

Quezaltenango is the most important city of the western highlands. A line of cities follows along at the juncture between the upper Pacific coastal plain and the line of volcanoes. The largest of these are Retalhuleu, Mazatenango, and Escuintla. Puerto Barrios, Zacapa, and Cobán in the east and Huehuetenango to the west are the major urban centres in the sierras.

Rural settlements tend to radiate around the county seats (*cabeceras*) of the hundreds of *municipios* into which the country is divided. The living conditions in the vast majority of these contrast sharply with the modern amenities of Guatemala City. Running water and up-to-date sanitary facilities are lacking in most homes. Dwellings tend to be made of adobe, cane, or planks and roofed with thatch, tiles, shakes, or corrugated metal. Homes commonly have earthen floors.

The people. *Ethnic composition.* On the basis of cultural traits, the population is divided into two main ethnic groups—Ladinos and Amerindians. The Ladinos comprise those of mixed Hispanic-Amerindian origin in addition to those Amerindians who have adopted Western ways of life. While Amerindians account for some 45 percent of the nation's total population, they make up 75 percent of the population in the western highland provinces. The Ladinos are the more commercially and politically influential group, and they make up most of the urban population. Most of the small number of blacks, called Black Caribs, inhabit the Caribbean lowlands.

Amerindian languages

Linguistic composition. All official transactions in Guatemala are conducted in Spanish. There are, however, about 20 Amerindian languages spoken, most of them of Mayan origin. The largest Amerindian groups are the Mam, who reside in the western departments of Guatemala; the Quiché, who occupy areas to the north and west of Lake Atitlán; the Cakchiquel, who extend from the eastern shores of Lake Atitlán to Guatemala City; and the Kekchi, who are concentrated in the sierras to the north and west of Lake Izabal. Many of the Amerindians also speak Spanish.

Religion. While Roman Catholicism is the dominant religion of Guatemalans, among Amerindians it is often heavily infused with beliefs of pre-Columbian origin. From the mid-20th century, however, there has been a surge of conversions to Protestantism, particularly among the poor—that religion offering strong encouragement toward self-improvement. The most important shrine in Central America is the Roman Catholic Black Christ of Esquipulas (named for the dark wood from which it is carved), located in eastern Guatemala.

Demographic trends. Guatemala has a relatively high rate of population growth. Within Central America, Guatemala has one of the highest birth, death, and fertility rates, in addition to a low life expectancy. Due to impoverishment, thousands of the rural poor in search of a livelihood migrate seasonally to the Pacific coastal plain to harvest crops, to the major urban centres, or to sparsely populated areas in the Petén to the north. Civil conflict in the rural areas occupied by Indians has also led to the mass flight of refugees into neighbouring Mexico, Belize, and the United States.

The economy. Guatemala is a developing country largely dependent upon traditional commercial crops as the basis of its market economy. Vigorous economic growth during the 1960s and '70s was followed by a severe economic downturn during the '80s. The government has attempted to revitalize the economy by fostering the diversification and expansion of nontraditional exports, and free trade zones have been established to encourage the expansion and decentralization of manufacturing. The collection of personal income taxes has improved significantly; the government, however, continues to rely upon revenue from other tax sources, such as customs duties, sales taxes, and excises on liquor and tobacco.

Although labour unions and student and peasant organizations have greater freedoms, leaders continue to suffer harassments and threats of death. In an effort to sustain the government, the United States has greatly increased economic and military aid since 1980. Negative trade balances and foreign indebtedness, however, continue to hamper the economy.

Resources. Surveys have long reported Guatemala to have a wealth of mineral resources; this potential has had only minimal fulfillment, however. Since 1981 several million barrels of petroleum have been extracted annually from oil wells located in the Petén. The export of concentrates of lead, zinc, and antimony is minimal. The export of nickel from the region of Lake Izabal was terminated in 1981.

Arable land

The agricultural resources of Guatemala are rich. Although rugged landscapes prevail in much of the volcanic region, numerous highland basins and the Pacific piedmont and coastal plain provide productive soils for agriculture. Within the sierras, the lower Motagua valley offers excellent soils. The wide range of climates allows for a diversity of crops. The efficient exploitation of soils is primarily limited by the inequitable distribution of land (large landowners not being required to maximize land use) and the inability to provide the agricultural sector with adequate financial support—*i.e.,* funding of small farms.

Forest and fishing resources have considerable potential. Forest products are derived primarily from the tropical forests of the Petén and the coniferous forests of the highlands. Limited accessibility, however, hinders the exploitation of forest resources. The coastal waters of the Pacific in particular provide fishing resources of some importance.

Agriculture, forestry, and fishing. Although agriculture provides employment for more than one-half of the work force, it contributes only one-fourth of the gross national product (GNP). Traditional peasant agriculture, focused upon the production of corn (maize), beans, and squashes for domestic consumption, is concentrated on small farms in the highlands. In contrast, commercial plantation agriculture, emphasizing the production of coffee, cotton, sugarcane, bananas, and cattle for foreign markets, is restricted to large estates on the Pacific piedmont and coastal plain and in the lower Motagua valley.

Guatemala has developed into a major world supplier of cardamom. Increasingly, peasants who have long produced grains and beans and tended sheep are turning to the production of nontraditional commodities—fruits, vegetables, flowers, and ornamental plants—destined for export and for rapidly growing urban markets within the country.

Lumber, primarily pine, is exported in small volume and used domestically in construction and the crafting of furniture. But the most important use of cut wood is for fuel, more than half going to that purpose. Commercial fishing in the Pacific has developed and includes a catch of crustaceans, especially shrimp, and such fish as tuna, snapper, and mackerel; most of the catch is exported.

Industry. Manufacturing grew rapidly between 1960 and 1980 but expanded more slowly during the 1980s. During the late 1980s it contributed 15 percent to the GNP. Mining and the production of energy contributed an additional 3 percent. Food processing and beverages, the processing of tobacco and sugar, publishing, the manufacture of textiles, clothing, cement, tires, and pharmaceuticals, and the refining of petroleum are primary industrial activities. Most of the industrial activity is concentrated in the environs of Guatemala City.

Energy

The primary sources of energy are petroleum, hydroelectricity, and fuelwood. Fossil fuels and hydroelectricity both contribute substantially to the nation's electricity requirements. Throughout the more densely populated regions, wooded areas provide firewood and charcoal for cooking, heating, the firing of ceramic ware, and the production of lime.

The growing service sector is the largest contributor to the GNP. Increasing emphasis is being placed upon tourism as a source of income and employment. Noteworthy archaeological ruins are located at Tikal in the Petén, Zacaleu on the outskirts of Huehuetenango, and Quirigua in the lower Motagua valley. Antigua Guatemala offers a wealth of colonial architecture. Volcanic landscapes and mountain valleys provide incomparable settings for villages occupied by colourfully attired Indians. Of renown is the Indian marketplace of Chichicastenango.

Finance. The government-controlled Bank of Guatemala is the note-issuing authority and oversees the country's banking system. It also handles all international accounts. A number of other public and private banks are in operation, and a stock exchange was established in Guatemala City in 1987. Guatemala's monetary unit is the quetzal.

Trade. The United States is Guatemala's primary trading partner in terms of both imports and exports. Other major trading partners are Germany, El Salvador, and Japan. The Central American Common Market, founded in 1960, and of which Guatemala is a member, fosters trade between Central American nations. Mexico and Venezuela are major suppliers of petroleum. Other leading imports include electrical machinery, transport equipment, paper, metal products, pharmaceutical and other chemical products, textiles, and food. The major export is coffee, followed by sugar, bananas, chemical products, cardamom, and cotton. The exports of vegetables, fresh fruits, flowers, and seafoods are of increasing importance.

Transportation. A network of highways, concentrated in the southern portion of the nation, is the major means of transport. Railroads have markedly diminished as a means of conveying freight and passengers. Commercial domestic flights within the country are basically limited to those between Guatemala City and the Petén.

Highways

Two primary highways extend east-west across Guatemala. The Inter-American Highway lies to the north of the southern chain of volcanoes. A Pacific coast highway lies to the south. These routes are linked by a number of roads that pass through the chain. The Pacific coastal plain is served by a number of paved highways that extend south from the primary coast highway. The primary north-south highway extends from San José on the Pacific to Puerto Barrios on the Atlantic, by way of Guatemala City. By far the greater number of passengers are carried by bus rather than by private car.

The primary Pacific coast highway and the north-south interoceanic highway are paralleled by the nationally owned railroad. At Zacapa, a rail line branches southeast to El Salvador.

Most of the foreign trade is handled through the Caribbean port of Santo Tomás de Castilla. New Pacific port facilities (Puerto Quetzal) are in operation at San José.

La Aurora International airport, located on the southern outskirts of Guatemala City, is served by international airlines providing service to points throughout the Western Hemisphere and in Europe. The privately operated national airline is Aviateca.

Administration and social conditions. *Government.* The constitution adopted in 1986 defines the country as a sovereign democratic republic and divides power among three governmental branches: legislative, executive, and judicial. Legislative power is delegated to a unicameral Congress elected every five years by direct, popular suffrage. Executive power is vested in the president and vice president, also elected every five years by popular vote. All citizens over the age of 18 are obliged to register and to participate in elections.

The Supreme Court, with at least nine justices, has jurisdiction over all the tribunals of the country. The justices are elected by Congress for terms of four years. The national territory is divided into *departamentos,* each headed by a governor appointed by the president. The departments in turn are divided into *municipios* (municipalities), which are governed by councils presided over by mayors, elected directly by popular ballot.

Broad guarantees are provided for the organization and functioning of political parties, except for the Communist Party and any other that is deemed to be dedicated to the overthrow of the democratic process. Only authorized political parties may nominate candidates for president, vice president, and Congress. Candidates for mayor and other municipal offices need not be nominated by political parties.

Political parties

There is a constant flux in the formation and demise of political parties. Those displaying the most continuity are the Revolutionary Party (Partido Revolucionario; PR), which has shifted from left to right in political orientation, the centrist Guatemalan Christian Democratic Party (Partido Democrácia Cristiana Guatemalteca; PDCG), the right-leaning Institutional Democratic Party (Partido Institucional Democrático; PID), and the right-wing National Liberation Movement (Movimiento de Liberación Nacional; MLN).

Although constitutionally outside of politics, the army nevertheless represents a powerful element in political struggles and has often controlled the government.

Education. Education is in theory free, secular, and compulsory through the primary school. Secondary schools train teachers, agricultural experts, industrial technicians, and candidates for universities. An enrollment of 65 percent of those eligible to attend primary schools declines to less than 20 percent for secondary schools. In rural areas, even many of those who have attended primary schools (usually only to the third grade) are functionally illiterate as adults. Impoverishment and a low premium paid upon education contributes to the lowest level of literacy to be found in Central America.

Universities are concentrated in the capital. The largest is the national University of San Carlos, founded in 1676. The private universities of Rafael Landívar, Del Valle, Francisco Marroquín, and Mariano Galvez have been founded since 1960.

Health and welfare. The inadequacy of medical and health services, particularly in rural areas, is reflected in the high rates of infant mortality. Inadequate sanitation and malnutrition are contributory factors. In larger communities the Ministry of Public Health and Social Assistance maintains hospitals that provide free care, and there are also numerous private hospitals. During the 1980s rural health centres staffed by personnel trained in preventive medicine were established in hundreds of localities in an effort to improve the health of rural inhabitants.

Since 1946 the Guatemalan Social Security Institute has provided medical insurance for public and private employees. The benefits cover accidents and common illness, as well as maternity care. The institute also maintains several hospitals.

Cultural life. *Daily life.* The nation's society is marked by pronounced extremes in the conduct of daily life. In the capital city families live much as they do in the cosmopolitan centres of Europe, whereas within an hour's drive of the capital are Indians whose patterns of daily life remain those of past centuries. The pattern of culture is characterized by sharp contrasts, whether it be in the language spoken or in matters pertaining to the household, cuisine, attire, or family affairs.

The nation's populace is increasingly exposed to the in-

Foreign cultural influences

trusion of foreign influences upon their way of life. All aspects of communication—periodical news, the comics, soap operas, film—are primarily of foreign origin. A multitude of products, from soaps and boxed cereals to automobiles and bottled drinks, bear foreign brand names.

The arts. The evidence of Mayan culture pervades the country. Today, although native crafts involve a variety of forms of expression, they are best represented in colourful handwoven textiles. Traditional dances, music, and religious rites that have survived in the Amerindian regions are important tourist attractions. The art of the colonial period is chiefly represented in the architecture and decor of Roman Catholic churches.

For the most part, the recognition of modern painters, composers, and authors tends to be limited to within the nation. A major exception is the author Miguel Angel Asturias, a poet and novelist who received the Nobel Prize for Literature in 1967. As with Asturias' writings, expressions of the various art forms tend to focus upon the Amerindian heritage of the nation.

Cultural institutions. Virtually all of the more highly recognized centres of cultural activity are concentrated within Guatemala City. These include the National Theatre, the Conservatory of Music, the Museum of Modern and Contemporary Arts, the Ixchel Museum of Indian Attire, a museum of natural history, and the National Museums of Archaeology and Ethnology and of Arts and Popular Crafts. The National Archives have a rich collection of materials on Mexico and all Central America except for Panama. The Society of Geography and History ranks as one of the oldest and most highly respected learned societies in Guatemala.

Recreation. Heavily attended fairs and religious festivals are scheduled in every part of Guatemala throughout the year. Easter celebrations at Antigua Guatemala draw enormous throngs. Among the most important of the national holidays is the day of independence, September 15. The media devotes much attention to soccer, a sport that is widely enjoyed in all parts of the nation. In larger communities, recreational parks draw large crowds on weekends.

Press and broadcasting. Major newspapers and publishing houses, as well as radio and television stations, are located within the capital. Among the most widely circulated newspapers are *La Prensa Libre, El Gráfico, La Hora,* and *Siglo Veintiuno. Diario de Centroamérica* is published by the government. Radio and television have assumed a major role in reaching large numbers who are illiterate or who reside in remote areas of the nation. All means of communication are ostensibly free of government censorship. Censorship has been imposed in times of crisis, however, and intimidation and threats of physical harm have often hindered the free expression of thought. (O.H.H.)

For statistical data on the land and people of Guatemala, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Origin of the republic's name

The origin of the name Guatemala is Indian, but its derivation and meaning are undetermined. Some hold that the original form was "Quauhtemallan" (indicating an Aztec rather than a Mayan origin), meaning "land of trees," and others hold that its origin was "Guhatezmalha," meaning "mountain of vomiting water," referring no doubt to such volcanic eruptions as the one that destroyed Santiago de los Caballeros de Guatemala, the first permanent Spanish capital of the captaincy general.

The colonial period. Under the Spanish, the capital was reestablished at the nearby location of present-day Antigua Guatemala. The capital achieved a certain magnificence, and the other major towns acquired some aspects of Spanish culture, but the outlying areas were only lightly affected. When the capital was razed by an earthquake in 1773, it was moved by royal order to the site of modern Guatemala City.

The colony developed no great degree of economic prosperity. The cultivation for export of agricultural staples, principally cacao and indigo, by Indian or black labour was the major economic activity, exclusive of production for subsistence. Commerce, however, was never extensive; a satisfactory port was never developed, internal transportation was difficult, and pirates harassed the coasts and preyed on shipping. The importance of Guatemala City lay in the fact that it was the administrative centre of the region, being the headquarters of the captain general and the high court (Audiencia).

The postcolonial period. Following independence from Spain (1821) and Mexico (1823), Guatemala was the political centre of the United Provinces of Central America. The principal factor in the collapse of the federation was the backcountry uprising in Guatemala led by Rafael Carrera, who established himself as the military arbiter of the state (1838) and, from the executive's chair or from behind it, controlled policy until his death in 1865. Elections were dispensed with in 1854, when the presidency was conferred upon him for life.

Carrera, with Conservative support, returned Guatemala to a regime similar to that of the colonial period. He restored the church to its position of privilege and power and catered to the aristocracy. Carrera followed a nationalistic policy, and in March 1847 he formally declared Guatemala an independent and sovereign nation. In 1859 he failed to get Britain to follow through on a treaty defining the status and boundaries of British Honduras, an issue that remained unsettled even after Belize became independent.

Revolution of 1871

In 1871 a revolution headed by Miguel García Granados and Justo Rufino Barrios overthrew General Vicente Cerna, Carrera's Conservative successor in office, and inaugurated a period of Liberal ascendancy that extended almost unbroken to 1944. After a brief period in the presidency, García Granados ceded to Barrios (1873), who became known as the Reformer because of the sweeping changes he introduced.

With the approval of the assembly, Barrios broke the power of the local aristocracy; brought the church under civil control and confiscated its properties; instituted lay education; promulgated a new constitution (1876); fostered the construction of roads, railways, and telegraph lines; encouraged development by private initiative of Guatemala's resources; and opened the country to foreign capital. He stimulated the cultivation of coffee and enacted legislation that assured producers of a ready supply of labour. He was an ardent exponent of a Central American union, and, when political means failed to produce results, he invaded El Salvador and lost his life at the Battle of Chalchuapa (1885) in an attempt to accomplish his goal.

After the death of Barrios, Manuel Lisandro Barillas occupied the presidency. He was succeeded by José María Reina Barrios, a nephew of the Reformer, who was elected in 1892 and assassinated in 1898. Manuel Estrada Cabrera then became provisional president, regularized his status by an election, and by repeated reelections maintained himself in power until leaders of the opposition Unionist Party forced him from office by having the assembly declare him insane (1920). During his long tenure, Estrada Cabrera fostered economic development and progress along the lines established by Barrios. He encouraged improvements in agriculture, made concessions to the United Fruit Company (owned by U.S. businessmen), continued to build roads, and supported railroad construction, seeing completion of the railroad to the Atlantic. Health conditions were improved, and education was stimulated. Estrada Cabrera persecuted political opponents, disregarded individual rights, muzzled the press, and summarily disposed of his enemies.

After the fall of Estrada Cabrera the presidency was held by a series of short-term rulers. In 1931 General Jorge Ubico was elected president without opposition and began the fourth of Guatemala's extended dictatorships.

Jorge Ubico

Contemporary period. Ubico stressed economic development and, in particular, the improvement and diversification of agriculture and the construction of roads. He balanced the national budget and transformed a deficit into a surplus. His paternalistic policies toward the Indians established him as their patron, although his vagrancy law (1934) made workers, especially Indians, liable to periods of forced labour at critical seasons. During his motorcycle tours of the country or in his office, he listened to their complaints and dispensed immediate "justice." This relationship deluded Ubico (called Tata, "Father") into stating that Guatemala no longer had an Indian problem. Education was of but slight interest to him.

Ubico's administration dramatized the degree to which Liberal thought had lost its idealism and was concerned principally with material progress. The new socioeconomic groups found no stimulation and no hope in the dreary materialism and military repression that had come to characterize Liberal regimes, and these potential sources of opposition were brought together by the increasing disregard shown for individual rights and liberties. The discontent was increased by economic dislocation during World War II. In December 1941 Ubico's government declared war on Japan, Germany, and Italy.

Juan José Arévalo

In June 1944 a general strike forced Ubico to resign. Labour was allowed to organize, political parties were formed, and a presidential electoral campaign was begun, in which Juan José Arévalo soon emerged as the most popular candidate. General Federico Ponce Vaides, head of the interim government, was deposed on Oct. 20, 1944, by a popular uprising; and a revolutionary junta presided over the drafting of a new constitution and the electoral campaign, which was won by Arévalo. The Arévalo administration attempted to consolidate the social revolution implicit in the October uprising. A favourable labour code was enacted, and a social security system that promised progressive extension of benefits was inaugurated. Arévalo also pressed the Belize issue with Great Britain, subjected foreign enterprises to regulation, and attempted to guarantee Guatemalan labourers larger benefits. Thus, the Arévalo regime transferred political power from the military to a popular group, of which organized labour was the most important element.

Lack of leadership from the rank and file allowed Guatemalan communists to organize the labour movement and use it for their own ends. Arévalo was not friendly to their activities, but his nationalistic bent gave them opportunity to establish themselves as his most enthusiastic and reliable supporters.

Successors to Arévalo

The most likely candidates to succeed Arévalo were two military men, Francisco Arana and Jacobo Arbenz. Arana was assassinated in 1949. Arbenz became the official candidate, was elected—with communist support—over General Miguel Ydígoras Fuentes, and assumed office in March 1951. Arbenz made agrarian reform the central project of his administration. With communist support, the National Congress passed a measure providing for the expropriation of unused portions of landholdings in excess of a specified acreage and for the distribution of the land among landless peasants.

The land reform, which had a heavy impact upon the U.S.-owned United Fruit Company, and the growth of communist influence became the most troublesome issues of the Arbenz regime. The U.S. Central Intelligence Agency (CIA) began efforts to destabilize the regime and recruited a force of Guatemalan exiles in Honduras, which was led by the exiled Colonel Carlos Castillo Armas. When the invasion began, the army would not fight for Arbenz, however, and in June 1954 he was forced to resign.

Castillo Armas emerged from a military junta as provisional president, and a plebiscite made his status official. He extirpated communist influence, quashed agrarian reform, and broke labour and peasant unions with considerable violence. He sought to restore the confidence of foreign investment capital, but he was assassinated in July 1957. After two temporary governments and an election nullified by the congress, Ydígoras Fuentes was declared elected to the presidency and took office on March 2, 1958. In March 1963 a coup led by the defense minister, General Enrique Peralta Azurdia, removed Ydígoras. As provisional president, Peralta canceled the election, promising to hold one later. The frustration, during these regimes, of the social reforms promised by the revolution of 1944 made restive elements of the population increasingly receptive to counsels of resistance and even of violence.

An orderly election on March 6, 1966, gave Julio César Méndez Montenegro, a law professor and the candidate of the moderately left-wing Revolutionary Party (Partido Revolucionario; PR), an unexpectedly large plurality of votes over the candidate of the military regime but not the absolute majority required for election. Congress elected him, but the understanding with the military officers that had to be reached before a civilian government could take office undermined his authority. Hopes for reform, therefore, were largely frustrated, and the energies of the administration were consumed in attempts to control the increasing violence and terrorism. Rural guerrillas were substantially eliminated by military and paramilitary operations such as those conducted by Colonel Carlos Arana Osorio, but urban guerrilla and terrorist activity worsened.

Carlos Arana Osorio

Arana Osorio, the "law-and-order" candidate, won the election of 1970. His major activity was "pacification" of the country by the extermination of "habitual criminals" and leftist guerrillas. Assassination of opposition leaders of the democratic left by so-called death squads, often linked to the military and the police, gave rise to the conviction that Arana was attempting to eliminate all opponents, whether left, right, or centre. With dissent eliminated or hushed, the country experienced a period of relative quiet. As the election of 1974 approached, optimists could find some reason to hope that a new basis had been laid for reform. The coalition of opposition parties chose General Efraín Ríos Montt, a leading officer of the progressive wing of the military forces, to contend with General Kjell Laugerud García, a nonpolitical military officer representing the coalition of rightist parties.

When returns showed Ríos Montt winning an absolute majority, which constituted election, the government abruptly suspended election reports, brazenly manipulated the results, and finally announced that Laugerud García had won a plurality of votes. The government-controlled National Congress promptly elected him. Deprived of moral force, Laugerud took office as the protégé of Arana. He faced problems of inflation, a series of volcanic eruptions, and division and consequent weakening of his principal political support, the MLN. He met a renewal of leftist violence and terror with the same repressive measures that Arana had applied.

The pattern of electoral manipulation set in 1974 persisted in subsequent elections. General Fernando Romeo Lucas García, declared the winner in 1978 after another suspect count, presided over a regime that essentially continued that of Laugerud. Both administrations confronted the nation's problems with resources greatly reduced in consequence of natural calamity. The devastating earthquake of February 1976 was labeled the greatest natural disaster ever recorded in Central America.

A major factor in both administrations was the discovery of oil in northern Guatemala. Because the deposit was thought to extend across British Honduras (now Belize) to the continental shelf, resolution of persistent, conflicting boundary and territorial claims was sought. On March 11, 1981, Guatemala, Great Britain, and British Honduras reached preliminary agreement, but a final settlement was not reached, and in September 1981 Great Britain granted independence to Belize over Guatemala's protest. The discovery of oil was also thought by some to be behind government violence in the largely Indian departments of the north. The devastation that occurred there drove thousands of Indians into Mexico, suggesting that the administration might be clearing lands for others to appropriate. Certainly, Indians moved in unprecedented numbers into the guerrilla movements.

In the elections of March 1982, the government coalition candidate, General Angel Aníbal Guevara, was declared elected. On March 23, however, young army officers seized the government and installed a junta headed by General Ríos Montt, who had been denied the presidency in 1974.

Ríos Montt

Ríos Montt dissolved the junta and pledged to rout corruption, disband the notorious death squads, and end the guerrilla war. The new leader failed to follow through on his promises, however, and conditions in Guatemala worsened. Ríos Montt's economic policies were not effective, and the political violence that he had promised to end was soon renewed with even greater intensity, again forcing many peasants to flee into Mexico and driving others into guerrilla camps, thus fueling the insurgency. A Protestant in a largely Roman Catholic country, Ríos Montt had trouble holding support.

In August 1983 Ríos Montt, unable to hold support, was overthrown by General Oscar Humberto Mejía Víctores, who promised a quick return to the democratic process. Violence continued in the countryside, however, and the United States, seeking human rights improvements, restricted aid to the new regime; military aid had been curtailed since 1977 for human rights violations. Elections for a constituent assembly were held in July 1984, and the parties of the centre pulled about a third of the vote, indicating a movement away from the support of the right-wing groups said to influence the kidnappings, murders, and other violence often attributed to death squads.

A new constitution, bringing greater emphasis to human rights guarantees, was approved in May 1985, and presidential elections held the following December produced a landslide victory for the centrist PDCG leader, Marco Vinicio Cerezo Arévalo, who received some 68 percent of the vote. It was the first election of a civilian president in Guatemala in 15 years. Despite Cerezo's sweeping victory, the new government faced onerous and complex problems that could only be resolved with consummate skill and effort. On taking office Cerezo told the nation to prepare itself for severe austerity measures that were necessary to combat Guatemala's worst economic depression in 50 years.

Continued reliance on U.S. aid

Hopes that Cerezo's election could support human rights reforms and end the violence were quickly dashed. There was a resurgence of death squad activity, particularly in the capital. The various bands of Marxist guerrillas, largely checked in the time of Ríos Montt and Mejía Víctores, found a new unity in the formation of the Guatemalan National Revolutionary Unity (Unidad Revolucionario Nacional Guatemalteco; URNG). A series of attempted military coups were put down by Defense Minister General Héctor Alejandro Gramajo. Labour and peasant unrest also increased during the Cerezo presidency. Some painful economic progress was made, but Cerezo had to rely on U.S. aid as the insurgency and violence continued to grow in intensity into the 1990s. (W.J.Gr./T.P.A.)

For later developments in the history of Guatemala, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Honduras

The Republic of Honduras (Spanish: República de Honduras) is a small nation of Central America situated between Guatemala and El Salvador to the west and Nicaragua to the south and east. The Caribbean Sea washes its northern coast, the Pacific Ocean its narrow coast to the south. With an area of 43,277 square miles (112,088 square kilometres)—including the offshore Caribbean department of the Bay Islands—Honduras is one of the largest countries of Central America. The bulk of the population lives a generally isolated existence in the mountainous interior, a fact that may help explain the rather insular policy of the country in relation to Latin and Central American affairs. Its economy is primarily agricultural, bananas and coffee for export being particularly important, although corn (maize) is the chief domestic staple. The capital is Tegucigalpa, but—unlike most other Central American countries—there are other cities, such as San Pedro Sula, that are equally important industrially and commercially.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* More than 75 percent of the land area is mountainous, lowland being found only along the coasts and in the several river valleys that penetrate toward the centre of the interior. The interior takes the form of a dissected upland with numerous small peaks. The main surface features have a general east-west orientation. Except for a narrow plain of alluvium bordering the Gulf of Fonseca in the south, the southwestern mountains, the Volcanic Highlands, consist of alternating layers of rock composed of dark, volcanic detritus and lava flows, both of Tertiary age (from 1.6 million to 66.4 million years old). The northern mountains in other regions are more ancient, granite and crystalline rocks predominating.

The four regions

Four distinct geographic regions may be discerned: (1) The eastern Caribbean lowlands (including the northern part of the area called the Mosquito [Miskito] Coast) and mountain slopes embrace about one-fifth of the total land area. Hot and humid, this area is densely forested in the interior highlands, lumbering being an important economic activity. Subsistence agriculture and fishing are the main support of the scattered population. (2) The northern coastal and alluvial plains and coastal sierras make up about one-eighth of the land area and contain about one-fourth of the population. This is an economically important area, the clayey and sandy loam soils producing rich crops of bananas, rice, cassava (manioc), oil palm, corn, citrus fruits, and beans. Cattle, poultry, and pigs are raised. The nation's railroads are confined to this northern area, which has four of the five important ports of entry. (3) The central highlands take up two-thirds of the national territory and contain the vast majority of the population. The mountains are rugged, rising to more than 9,200 feet (2,800 metres) in the west, and the numerous flat-floored valleys are situated at altitudes of from 1,000 to 3,000 feet. The generally fertile soils, derived from lava and volcanic ash, produce coffee, tobacco, wheat, corn, sorghum, beans, fruits, and vegetables. Cattle, poultry, and pigs are raised. (4) The Pacific lowlands, centred on the Gulf of Fonseca, and the adjacent lower mountain slopes constitute only a small part of the land area and contain an equally small part of the population. The soils are fertile—composed of alluvium or volcanic detritus—and produce sesame seed, cotton, and small amounts of corn and sorghum. Cattle are raised on the lowland pastures, and coffee is grown on the nearby uplands.

Climate. The climate is generally hot, with high humidity in the tropical coastal lowlands becoming modified by altitude toward the interior. Lowlands below 1,500 feet have mean annual temperatures between 79° and 82° F (26° and 28° C). The north coast is occasionally affected from October to April by cool northern winds of continental origin. Mountain basins and valleys, from 2,000 to 4,000 feet, have mean annual temperatures of 66° and 73° F (19° and 23° C). At Tegucigalpa, located on hilly terrain at an altitude of 3,200 feet, the rainy season starts in May and continues until the middle of November, with temperatures sometimes reaching 90° F (32° C) in May and dropping to 50° F (10° C) in December, the coolest month. Around 7,000 feet, mean annual temperatures are about 58° F (14° C). In the northern and eastern coastal and alluvial plains and on adjacent mountains, mean annual precipitation ranges from 70 to 110 inches (1,800 to 2,500 millimetres) or more, with a less rainy season from March to June; these areas occasionally have summer hurricanes with heavy rains. Pacific plains and mountain slopes get 60 to 80 inches (1,500 to 2,000 millimetres) of rain annually but from December to April receive little or no rain. Interior sheltered mountain basins and valleys receive 40 to 70 inches annually.

Coastal and lagoon swamps

Plant and animal life. In eastern Honduras the coastal and lagoon swamps have mangrove and palm forests, and west of these are low, rainy, sandy plains with pine-savanna (*Pinus caribaea*) forests, extending inland for 40 miles or more. West of the pine-savanna forests, in low valleys and on lower mountains, which are rainy all year, and on the low, rainy northern mountains are broad belts of dense evergreen broad-leaved forests with many species of large trees, including mahogany, lignum vitae, Spanish cedar, balsa, rosewood, ceiba, sapodilla (yielding chicle used for the base of chewing gum), and castilloa rubber. The high, rainy mountain slopes of highland Honduras support excellent oak-pine forests. Open, dry, deciduous woodlands and temperate grasslands are spread throughout the interior highland basins and valleys. The Pacific plains and adjacent mountain slopes have deciduous tropical forest and savanna. Mangrove occupies the low coastal swamps.

Insects, birds, and reptiles are the most conspicuous animal forms. There are many species of butterflies, moths, beetles, bees, wasps, spiders, ants, flies, and mosquitoes, many of them beautifully coloured. Waterfowl in large numbers inhabit the coastal areas. Crocodiles, snakes, lizards (giant iguana and others), and turtles are found in the tropical forest areas. The fauna also includes deer, peccaries, tapir, pumas, jaguars, and ocelots. Fish and mollusks are abundant in lagoons and coastal waters.

The people. Honduras has been inhabited since well before the Common Era: the ruins at Copán in western Honduras indicate that the area was the centre of Mayan civilization before the Maya migrated to the Yucatán Peninsula. Most of the American Indians are Lenca and are now found in the southwest, near the Guatemala border, close to the most important Indian centres of the pre-Columbian period. Small, isolated groups of non-Spanish-speaking Indians—such as the Jicaque, Miskito (Mosquito), and Paya—continue to live in the northeast, although their numbers are declining. The largest sector of the population, however, is mestizo (a mixture of Spanish and Indian). The mestizos of Honduras are generally darker and more Indian in appearance than those of other Latin-American countries such as Chile, Uruguay, and Argentina. Blacks of West Indian origin and Black Caribs make up a small part of the population, living mainly along the Caribbean coast, an area in which considerable English is spoken.

Official language

The official language of Honduras is Spanish, and the predominant religion is Roman Catholicism, more than 90 percent of the population being adherents. The largest of the remaining groups are Protestant, with notable congregations in the east and on the Bay Islands.

A pronounced shift in population took place during the early part of the 20th century from the interior to the hot, humid north coast, where employment opportunities were provided by the United Fruit Company. These northwestern lowlands and the western and southern highlands constitute the most densely populated parts of the country. The population is growing extremely fast, posing a considerable problem in employment and housing. The majority of the population is rural, living in small villages or isolated settlements. The mountainous, forested terrain and rather poor road transportation are conducive to local isolation.

The economy. Honduras is a poor country, and the majority of Hondurans work under extremely difficult conditions. The government has, however, adopted more active economic policies since the mid-1900s. In 1954 the trade union movement obtained one of its most resounding triumphs, resulting in the promulgation of a labour code that is considered one of the most complete instruments of its kind in Latin America. The code has generally resulted in a higher standard of living for the worker and better operating conditions for business.

Honduras possesses a limited natural resource base that includes agricultural lands along the northern coast and interior river valleys, extensive pine forests not yet exploited, and small deposits of silver, lead, zinc, and low-grade iron ore. The economy is geographically divided between the highlands, where subsistence farming, stock raising, and mining have long dominated, and the lowlands, where plantation agriculture based largely on bananas is the chief occupation.

Agriculture, forestry, and fishing. In terms of the gross domestic product, agriculture contributes the largest portion. Agriculture also employs the majority of the labour force. Two large U.S. corporations—United Brands (formerly United Fruit Company) and Standard Fruit and Steamship Company—hold a disproportionate amount of the country's agricultural land and produce a substantial part of the national income by growing the majority of the country's banana crop. Important export crops other than bananas include coffee beans, cotton, tobacco, and sugarcane. Cattle raising is the main livestock activity, and beef has become an important export.

About two-thirds of the country's land is covered by forests, making forest products a potentially large source of national income. The extensive pine forests were attacked by blight in the 1960s, and mahogany—the major timber export—began declining in importance. The practice of shifting agriculture, employing widespread burning of forests and the cutting of wood for fuel, has caused a depletion of the forest resources. Present commercial practices of forest exploitation are inefficient. A substantial portion of timber harvested for commercial purposes does not reach the sawmill, and less than half of the timber that arrives at the mill is processed into lumber. To help alleviate the wasteful forestry practices the government put all forest trees under state ownership in 1974, but forests continue to be depleted at a rapid rate.

Depletion of forest resources

Fishing is a small but developing industry, carried on mainly off the Caribbean coast. Shrimp is the most important part of the catch, the largest portion of which is shipped to the United States.

Industry. Manufacturing is dominated by small-scale firms that are at intermediate levels of technology and possess limited processing capabilities. The major products manufactured and processed are food products, beverages, textiles, clothing, chemicals, lumber, and paper products. The production of capital and heavy intermediate goods is minimal. Industrial plants are located largely in the urban areas of San Pedro Sula and Tegucigalpa.

Mineral resources are considerable, though largely unexploited, and include antimony, iron, mercury, and copper. Production of silver and gold has remained of some importance, and lead and zinc are produced in commercial quantities.

Honduras has a few hundred miles of public-service railroads. The Honduras National Railway, which is owned by the government, extends from Puerto Cortés to San Pedro Sula. It hauls timber and agricultural products. The Tela Railway, once owned by the United Brands Company and acquired by the Honduras government in 1975, provides service for plantations in the eastern Sula Valley and the coastal plain. Another government-owned railroad runs east along the coastal plain to Balfate, with a branch extending into the Aguan Valley.

Roads in Honduras carry the vast majority of the freight tonnage and almost all the passenger traffic. The rugged terrain of the Honduras countryside favours the development of roads instead of railroads. The exception to this is along the northern coastal plain where railways serve the banana plantations. The heart of the primary road network is the north-south highway that links the Pacific port of San Lorenzo and the Caribbean port of Puerto Cortés, stopping at Tegucigalpa and San Pedro Sula. The highway provides access to important agriculture areas in the Sula and Choluteca valleys and along the Caribbean coast. The Inter-American Highway cuts across southern Honduras for about 100 miles. Highways also run southwest from San Pedro Sula to the El Salvador border and along the north coast from San Pedro Sula to La Ceiba.

All Honduran ports are operated by the National Port Authority. The major ports in the country are Puerto Cortés, Tela, La Ceiba, and Puerto Castilla. The Pacific coast provides deepwater anchorage at Amapala on El Tigre Island and at the mainland port of San Lorenzo, completed in 1978.

Domestic air travel, although declining, supplements rail and highway travel, which includes intercity bus and truck service. Villeda Morales Airport at San Pedro Sula is the largest airfield. The airport at Tegucigalpa has a substantially shorter runway and is minimally suitable for modern jet travel.

Administration and social conditions. *Government.* Since acquiring independence in 1821, Honduras has been described constitutionally as a democratic, representative, unitary state with power divided among the legislative, executive, and judicial branches. The country's constitution was rewritten 17 times between the years 1821 and 1982. Power, however, has frequently been mobilized and changed by violent, undemocratic means. Although the legislature is given the power to pass laws, practically all important legislation is drafted by the executive and his assistants. The National Assembly in theory has great authority to check the administrative activities of the president, but only during the period 1925–31, when several Cabinet ministers appointed by the president were forced to resign through censure, was such authority effective.

Power of the executive

The president is elected directly by popular vote for an unrenewable term of four years. The single-house National Assembly is composed of 134 legislators elected to four-year terms. The major political parties are the Liberal Party of Honduras (Partido Liberal de Honduras) and the National Party (Partido Nacional). All citizens over 18 years of age are permitted to vote.

For purposes of local administration, Honduras is divided into 18 *departamentos*. Governors are appointed by the president, one for each department, to carry out central government decisions. The departments are divided into *municipios* (municipalities), which are further partitioned into *aldeas* (villages, or hamlets). Rural areas are grouped into *caseríos* (settlements), which are subdivisions of *aldeas*. Localities may elect a mayor, a legal representative, and a council.

The justices of the Supreme Court are appointed by the president. The Supreme Court exercises centralized control over the lower courts, including the appointment of justices, and has original and exclusive jurisdiction to declare acts of the legislature unconstitutional.

Education. The Honduran educational system follows the European model of centralized control through the Ministry of Public Education. According to law, education is free and, at the primary level, compulsory for all children. Efforts have been made to combat illiteracy, which is very high for older people. Higher education is centred at the National Autonomous University of Honduras in Tegucigalpa (founded 1847).

Welfare and health. By the late 20th century Honduras, like most other Latin-American countries, had moved more and more in the direction of the interventionist, or welfare, state. In 1955 the Honduran basic labour code came into effect, granting the right to work, minimum wages, an eight-hour day, freedom to form labour unions, collective bargaining, conciliation, and the right to strike. Social security and social welfare benefits were not improved appreciably, however, and many Honduran workers outside the public sector and not employed by business or industry are not covered. Health care is also generally inadequate for the poor urban and rural labourers. Death rates are high among the lower economic groups, who suffer from two severe health problems in particular, malnutrition and malaria.

The Honduran basic labour code

Cultural life. The art and architecture of pre-Columbian and Spanish colonial periods are strongly evident in Honduran culture. Of special interest is the great Mayan city of Copán, which was restored in 1839 after its discovery in a thick jungle more than 250 years earlier; it represents the height of the Mayan Classic period. Spanish architecture reflects Moorish, Gothic, and, especially, Baroque styles. Modern Honduran culture has not produced many strong representatives of its art, the country's widespread poverty being a major impediment. Most contemporary artists reflect their colonial heritage, while the pre-Columbian heritage is seen mainly in Indian crafts. Social themes may also be reflected in paintings and literary works, the latter being generally represented by poetry and short fiction.

Cultural institutions in Honduras include the National School of Music and the National Museum, both in Tegucigalpa, and the Archaeological and Historical Museum in Comayagua. Noteworthy among other institutions are those dedicated to the production of theatrical works in both Spanish and English.

Family recreation often revolves around religious festivals honouring local saints. Soccer (football) is the favourite team and spectator sport, and international matches often arouse great emotion.

There is general freedom of the press in Honduras, and daily newspapers are published in the principal cities of the country. Those of Tegucigalpa and San Pedro Sula are most noteworthy. The progressive and rapid development of radio and television has provided the country with excellent facilities for speedy and effective communication. There are private radio and television networks that cover the entire country. (J.R.M.R./Ed.)

For statistical data on the land and people of Honduras, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Early history. When the Spanish arrived to colonize Honduras, the land was occupied by a variety of indigenous peoples, the most advanced of whom were the Maya. During the early colonial period, in the 1570s, a silver strike in the highlands brought a rush of prospectors to Honduras, resulting in the rise of an important population centre at Tegucigalpa, which competed thereafter with the traditional capital, Comayagua. Agriculture, however, the enduring economic base of Central America, was slow to develop in Honduras. Development of Spanish society in the Honduras area was hindered by coastal attacks from the pirates and buccaneers endemic to the Caribbean Sea and eventually by a concerted British effort to control the coastal areas of Central America. For long periods the Spanish utilized a soft defense against the Caribbean threat, falling back to the highlands and to the Pacific coastal areas, which were generally closer to their network of communication and transportation. Thus, the British came to control the Caribbean's Mosquito coastal region. The Sambo-Miskito peoples along the coast were the indispensable allies of the British in this endeavour. In the 18th century, however, the Spanish Bourbon kings made a sustained effort to recover the Caribbean coastal areas, and their success in the Gulf of Honduras was manifested by the completion of a fort at Omoa on the gulf by 1779.

Indigenous peoples

Independence from Spain came after 1821. Francisco Morazán, elected to head the United Provinces of Central America in 1830, was a Honduran Liberal whose influence derived from a principal goal of liberalism, to curtail the traditional power and privileges of the church and clergy. Honduras declared its absolute independence on Nov. 5, 1838. The pro-church Conservatives in Honduras took control under Francisco Ferrera, who was inaugurated as the first constitutional president on Jan. 1, 1841.

Domination by the Conservatives lasted until the 1870s, and the church was restored to its former position. In 1861 the Honduran government signed a concordat with the Holy See. After 1871 the ascendancy of Justo Rufino Barrios in Guatemala influenced a return to liberalism in Honduras, where Marco Aurelio Soto, a Liberal, assumed the presidency (1876). A new constitution (1880) sought to undo the work of the Conservatives. Five years later, Liberals in Honduras and elsewhere proved to be nationalists first and blocked an attempt by Guatemala to unify the isthmus by force.

The 20th century. In the first decade of the 20th century, the strongman in Nicaragua, José Santos Zelaya, put Miguel Dávila into the Honduran presidency. This led in 1911 and 1912 to something more serious than periodic revolutions. The U.S. president, William Howard Taft, sent marines to protect U.S. banana investments, which by this time had grown considerably, with three companies exploiting this Honduran product. All three made large capital outlays in the form of improved port facilities, railroads, workers' settlements, and similar developments.

In 1918 Honduras declared war on Germany but took no active part in World War I. Political upheavals continued after the war. In 1932, after years of unrest, another strongman, General Tiburcio Carías Andino, was elected president and remained in office until 1949, dealing with several abortive revolutions during his virtual dictatorship. Honduras declared war on Japan, Germany, and Italy in December 1941. Lack of shipping brought much economic distress during the war; export surpluses of bananas, coconuts, and copra piled up, leading to widespread unemployment and consequent unrest. But the government was able to maintain itself, and it promulgated some beneficial reforms. Carías survived a revolution in 1947, but a year later he resigned under pressure.

Carías Andino

Policies under Juan Manuel Gálvez (1949–54), Julio Lozano Díaz (1954–56), and Ramón Villeda Morales (1957–63) brought some modernization to the transportation system and to labour legislation. The International Court of Justice awarded the large disputed area north of the Coco River to Honduras in 1960. In 1963 Villeda Morales was deposed by a military coup led by Colonel Osvaldo López Arellano, who declared himself head of state. In the summer of 1969 the Soccer War with El

Salvador broke out, triggered indeed by a soccer game but caused by severe economic and demographic problems. Though brief, the war dampened hopes for economic and political integration in Central America.

After 1963 Honduras was ruled by military governments until the election of Ramón Ernesto Cruz (1971–72). Cruz's election resulted from the Soccer War, which Honduras had lost militarily. But López Arellano, as chief of the armed forces, retained real power, and in December 1972 he removed Cruz from office. Pressured toward modernizing reforms by younger military officers, López Arellano astonished many by announcing, in January 1974, a reform program that included land redistribution. His program had little success, however.

López Arellano was discredited and forced to resign in 1975 because of an international bribery scandal; he was replaced by Colonel Juan Alberto Melgar Castro (1975–78). Honduras prospered under Melgar. High earnings from the elevated world coffee market during those years were particularly helpful. His administration was weakened, however, by a series of scandals.

General Policarpo Paz García, who attained power through a bloodless military coup in late 1978, pledged to continue Melgar's policies, but he soon faced harder times. Central America entered a cycle of violence with the revolution in Nicaragua that overthrew Anastasio Somoza Debayle in the summer of 1979 and the revolution in El Salvador that was under way in that same year. Honduras appeared to be an island of stability as its neighbours experienced guerrilla warfare. In November 1981 the country elected a civilian government after 17 years of almost continuous military rule. The new Honduran president, Roberto Suazo Córdova of the Liberal Party, was a noted anticommunist who favoured strong relations with the United States. Hopes ran high for internal improvements, but these were dashed as Honduras became embroiled in the growing regional conflicts. Protests grew over the presence of Nicaraguan Contras, who were using U.S.-sanctioned Honduran border areas as bases for attacks against Nicaragua's Sandinista government. There was also dissension over U.S.-run camps for training Salvadorans in counterinsurgency to combat the growing civil war in their country; a ban against this practice was passed in 1984. These events seemed a reprise of Honduras' role in Central America during most of the 19th and 20th centuries, when, because of geography and a lack of national organization, Honduras had suffered repeated invasions and served numerous third parties as a base for attacks upon its neighbours.

Controversy over the Contras

The United States had hoped that its relations with Honduras would help establish the country as a model Central American democracy, but that image was tarnished when a controversy arose over Suazo's alleged manipulations in choosing the Liberal Party's presidential candidate for the November 1985 elections. It was finally agreed that each party could run multiple candidates, with the presidency going to the leading candidate of the party drawing the most votes. On that basis José Azcona Hoyo of the Liberal Party was declared the winner with only 27.5 percent of the vote, against Rafael Leonardo Callejas of the National Party, who drew about 41 percent. In the late 1980s the government joined the other isthmian governments in a cooperative movement for peace in Central America; this brought increased pressure at home and from abroad to restrict Contra activity and to reduce the U.S. presence. In 1990 a new president, Rafael Leonardo Callejas, took office; it was the first time in 57 years that an opposition government had taken office peacefully. (W.M.Cl.)

For later developments in the history of Honduras, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Nicaragua

The Republic of Nicaragua (Spanish: República de Nicaragua) is located in the middle of Central America and is bounded by Honduras on the north, the Caribbean Sea on the east, Costa Rica on the south, and the Pacific Ocean on the west. Despite territorial losses to its neighbours, it is the largest of the Central American republics, with an area of 50,464 square miles (130,700 square kilometres). Its population, mostly mestizo, represents a racial mixture that reflects the country's history. The capital is Managua, the nation's largest city.

Nicaragua can be characterized by its agricultural economy, its history of autocratic government, and its imbalance of regional development; almost all settlement and economic activities are concentrated in the western half of the country. Most of the population is engaged in subsistence farming, and the national economy is heavily dependent upon exports of coffee and cotton. The government was in the hands of the family of the former president Anastasio Somoza García from 1936 to 1979, when the Sandinista National Liberation Front seized power. The Sandinista government held power until 1990, when it was defeated in a national election.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* The western half of the country is made up generally of valleys separated by low but rugged mountains and many volcanoes. This intricately dissected region includes the Cordillera Entre Ríos, on the Honduras border, the Cordilleras Isabelia and Dariense, in the north-central area, and the Huapí, Amerrique, and Yolaina mountains, in the southeast. The mountains are highest in the north, and Mogotón Peak (6,900 feet [2,103 metres]), in the Cordillera Entre Ríos, is the highest point in the country.

Western mountains

To the west and south of the central mountain core is a string of about 40 volcanoes that stretches northwest-southeast along the Pacific coast. They are surrounded by low plains extending from the Gulf of Fonseca in the north to the Bay of Salinas in the south and are separated from the mountains by the great basin that contains Lakes Nicaragua, Managua, and Masaya. The volcanoes, some of which are active, are divided into two groups; in the north they are known as the Cordillera de los Marrabios and in the south as the Pueblos Mesas. The highest volcanoes include San Cristóbal, which rises to 5,840 feet; Concepción, which attains 5,108 feet; and Momotombo, at 4,462 feet.

The eastern half of Nicaragua has low, level plains. Among the widest Caribbean lowlands in Central America, they average 60 miles (100 kilometres) in width. The coast is broken by river mouths and deltas and large coastal lagoons and by coral reefs, islands, cays, and banks that dot Central America's largest continental shelf.

Drainage and soils. The central mountains form the country's main watershed. The rivers that flow to the west empty into the Pacific Ocean or Lakes Managua and Nicaragua. They are short and carry a small volume of water; the most important are the Negro and Estero Real rivers, which empty into the Gulf of Fonseca, and the Tamarindo River, which flows into the Pacific.

The eastern rivers are of greater length. The 425-mile-long Coco River flows for 295 miles along the Nicaragua-Honduras border and empties into the Caribbean on the extreme northern coast. The Río Grande de Matagalpa flows for 267 miles from the Cordillera Dariense eastward across the lowlands to empty into the Caribbean north of Pearl Lagoon (Laguna de Perlas) on the central coast. In the extreme south, the San Juan River flows for 124 miles from Lake Nicaragua into the Caribbean in the northern corner of Costa Rica. Other rivers of the Caribbean watershed include the 158-mile-long Prinzapolka River, the 55-mile-long Escondido River, the 60-mile-long Indio River, and the 37-mile-long Maíz River.

The west is a region of lakes. Lake Nicaragua, with an area of 3,156 square miles (8,157 square kilometres), is the largest lake in Central America. Located in the southern isthmus, the lake and its distributary, the San Juan River, have long been discussed as a possible canal route between the Caribbean and the Pacific.

Nicaragua's lakes

There are six freshwater lakes near the city of Managua. They include Lake Managua, which covers an area of 400 square miles; Lake Asososca, which acts as the city's reservoir of drinking water; and Lake Jiloá, which is slightly alkaline and is a favourite bathing resort. Lake Masaya is prized for its swimming and fishing facilities; the sulfurous waters of Lake Nejapa have medicinal properties ascribed to them; and Lake Tiscapa, which lies within the capital city, is the site of the Presidential Palace.

Other lakes in the Pacific watershed include Lake Apoyo, near Lake Masaya; Lake Apoyeque, picturesquely located between two peaks on Chiltepe Point, which juts into Lake Managua; and the artificial Apanás Reservoir on the Tuma River, which generates much of the electricity consumed in the Pacific zone.

Soils on the Caribbean coast are varied and include fertile alluvial types along waterways and relatively infertile types in the pine-savanna and rain forest regions. On the Pacific coast the soil is volcanic, and about 85 percent of its area is fertile.

Climate. The climate is slightly cooler and much wetter in the east than in the west. The Pacific side is characterized by a rainy season from May to November and a dry season from December to April. The annual average temperature is 81° F (27° C), and precipitation averages 75 inches (1,910 millimetres) yearly. On the Caribbean side of the country, the rainy season lasts for about nine months of the year, and a dry season extends from March through May. The annual average temperature is 79° F (26° C), and annual precipitation averages almost 150 inches (3,810 millimetres). In the northern mountains temperatures are cooler and average about 64° F (18° C). Prevailing winds are from the northeast and are cool on the high plateau, warm and humid in the lowland.

Tropical forest

Plant and animal life. More than one-third of the country is covered with a broad expanse of tropical forest. The numerous types of trees include the valuable cedar and mahogany. The quebracho, or axbreaker, yields hardwood; the guaiacum, or lignum vitae, yields one of the ironwoods; the guapinol yields resin; and the medlar produces fruit similar to the crab apple. The eastern lowlands comprise pine savannas, the streams of which are lined with forests.

The country is rich in wildlife. Deer are common in the forested regions, and the fauna also includes such cats as the puma, jaguar, ocelot, and margay; several species of monkey; and peccaries. The hot and rainy regions are inhabited by various types of reptiles, including crocodiles, lizards, snakes, and turtles. There are many species of water and land birds, and the rivers, lakes, and lagoons are inhabited by freshwater and saltwater fishes and mollusks. Rodents and insects are abundant.

Settlement patterns. The western volcanic mountains and surrounding lowlands and lakes contain the majority of the country's population, most of its cities, and most of its industry. The area also yields most of Nicaragua's agricultural produce. The valleys of the western central mountains contain a substantial population and yield about one-quarter of the national agricultural production. Although they make up about one-half of the country, the eastern lowlands are lightly populated. The region is important, however, as a vast resource of timber.

Nicaragua is predominantly urban. By far the largest city is Managua, on the southeastern shore of Lake Managua. Other important urban centres include León, Granada, Masaya, and Chinandega, all in the west. Matalgalpa, Estelí, Juigalpa, and Jinotega are among the largest cities of the central mountains. Bluefields and Puerto Cabezas are the largest towns on the Caribbean coast.

The people. Most of Nicaragua's population is a product of racial admixture. The overwhelming majority of the people are mestizos (persons descended from mixed European and American Indian ancestry), called Ladinos in Nicaragua. The country has about equal numbers of blacks and whites, together making up some 20 percent of the population, and American Indian peoples make up less than 5 percent of the total. The west coast has a small number of Monimbo and Subtiaba Indian peoples. The peoples of the east coast include Miskito, Sumo, and Rama Indians and Garifuna (descendants of African slaves and Carib Indians), and there are a number of Creoles (mostly black English-speaking peoples) and some Ladinos.

Languages

Nicaragua has many languages, but Spanish, the official language, is spoken by the vast majority of the people. American Indian languages are commonly spoken on the east coast, as is English. On the west coast Indian languages have all but disappeared from use, but their influence remains in place-names and in many nouns in Nicaraguan Spanish.

Roman Catholicism is the predominant religion of the Nicaraguan people. Small Protestant and Jewish communities are concentrated in the larger cities. There is no official religion.

Nicaragua's population was decimated by the Sandinista-Contra war that took place during the 1980s; the number of refugees and exiles leaving the country was estimated at some 20 percent of the population. As in the other countries of the region, the migrations within Nicaragua have been from rural to urban, causing a rapid growth rate in the cities.

The economy. By the 1990s Nicaragua had become one of the poorest of Latin-American countries as a result of almost two decades of civil strife, complicated by a U.S. economic embargo and failed Sandinista economic programs during the 1980s. By the time a newly elected government took over in 1990 the country suffered from massive shortages of goods, and economic assistance from countries such as Cuba and the Soviet Union had been greatly reduced. In addition, a huge foreign debt had accumulated. Aid from the United States and European countries became an important factor in dealing with the problems of rebuilding the economy.

Much of the economic sector was nationalized under the Sandinistas. The government in 1990 began to return to private control many economic activities, such as export industries. Public participation in the economy increased. Independent trade unions, having been limited under the Sandinistas, began again to add to their memberships.

Resources. Nicaragua is rich in natural resources, most of which have not been exploited on a large scale. The economy is basically agricultural, and industry is in an incipient stage of development.

Mineral resources include known deposits of gold, silver, zinc, copper, iron ore, lead, and gypsum. The forests are a vast resource of hardwoods and softwoods, and the inland and coastal waters contain abundant food fishes. Nicaragua's potential hydroelectric resources are considered immense.

Major crops

Agriculture, forestry, and fishing. Agriculture, forestry, and fishing engage as much as two-fifths of the labour force and produce about one-fourth of the total national income. The major cash crops are coffee and cotton, most of which are produced for export. Other crops include corn (maize), sugarcane, sorghum, bananas, rice, beans, plantains, cottonseed, sesame, tobacco, cassava (manioc), and wheat. A variety of fruits and vegetables are also produced for the local market.

Cattle are significant as a source of hides, meat, and dairy products in the west and of meat in the east. The cattle industry increased rapidly after World War II until the

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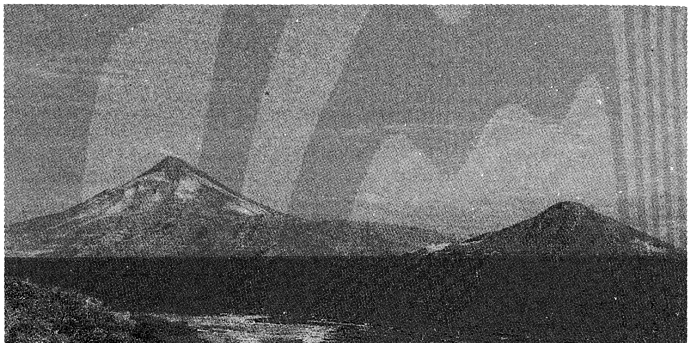

Momotombo Volcano (left) and Momotombito Island, viewed across Lake Managua, Nicaragua.

1980s, when internal conflicts and disruption prompted many ranchers to reduce or move their herds to neighbouring countries. Other livestock includes goats, hogs, horses, and sheep.

Forestry, once a rapidly expanding activity, also was reduced by the conflict of the 1980s, which destroyed most of the sawmills. Shrimping is the most important marine activity. Almost all the shrimps, caught in both the Pacific and Caribbean, are exported; lobsters are also exported in moderate quantities. Nicaragua's fish resources, however, are relatively unexploited because of lack of investment, and marine fishing remains largely a subsistence activity.

Industry. Of all the country's minerals, only gold was mined intensively. Mineral reserves are not exploited, because of lack of financing.

The industrial sector

Nicaragua's developing industry is based on the production of consumer products, many of which require the importation of raw materials. In the late 20th century the government actively supported the diversification of production and the use of domestic raw materials. Products include refined petroleum, matches, footwear, soap and vegetable oils, cement, alcoholic beverages, and textiles.

Finance. The Central Bank of Nicaragua, established in 1961, has the sole right of issue of the national currency, the cordoba. The financial system is dominated by the government-owned Finance Corporation of Nicaragua, an amalgamation of the nation's banks that was established in 1980.

Under the Sandinistas (1979–90), Nicaragua's chief trading partners were Cuba, the Soviet Union, and eastern European countries; a U.S. embargo limited trade to many other nations. In 1990, the new government reestablished trade relationships with Western countries. Imports are largely machinery, transport equipment, medicines, nonferrous metals, and petroleum. Exports include coffee, cotton, beef, bananas, gold, sugar, shrimps, and lobsters.

Transportation. Most of the country's transportation system is confined to the western zone. There is a network of highways, parts of which are impassable during the rainy season. The system includes the 255-mile (410-kilometre) Nicaraguan section of the Pan-American Highway, which runs through the west from Honduras to Costa Rica. An important road runs from the Pan-American Highway, 24 miles from Managua eastward to Port Esperanza at Rama. Another road connects Managua with Puerto Cabezas on the Caribbean.

There are a few hundred miles of railways. The main line runs from Granada northwest to Corinto, on the Pacific Ocean. A branch line leads north from León to the coffee area of Carazo.

The chief ocean ports of Corinto, which handles most foreign trade, Puerto Sandino, and San Juan del Sur serve the Pacific coastal area. The Caribbean ports include Puerto Cabezas and Bluefields, the latter connected to the river landing of Port Esperanza by regular small craft service. The short rivers in the west are navigable for small craft. In the east, the Coco River is navigable in its lower course for medium-sized vessels.

The main international airport, seven miles from Managua, has service to North America and Latin America. Another large commercial airport is at Puerto Cabezas. Other airports have scheduled domestic flights. Air services are offered by Aeronica, a private airline, and by other Central American and international airlines.

Administration and social conditions. *Government.* Nicaragua had 10 constitutions between 1838 and 1974. In 1979 a junta assumed power, supported by the Marxist-oriented Sandinista National Liberation Front (Frente Sandinista de Liberación Nacional; FSLN). It replaced the long dictatorial reign of the Somoza family, which had lasted 43 years. The old constitution was abrogated. An elected president and National Assembly replaced the junta and its appointed council in 1985, and a new constitution was promulgated in 1987.

The FSLN

The Sandinista regime lasted for almost 11 years, during which time there was virtually constant insurgency from as many as 13 resistance groups, some of which were known as the Contras and backed by the administration of U.S. President Ronald Reagan. In the election of February 1990 the Sandinista government was defeated by a 14-party coalition, the National Opposition Union (Unión Nacional Oposicion; UNO), and a new government was installed in April of that year. Immediate steps were taken to dismantle the socialist-style institutions of the Sandinista government and to return to a more typically Western-type democracy and a free-market economy. The Sandinistas retained enough seats in the National Assembly, however, to prevent constitutional change, impeding somewhat the progress of the new administration.

Education, health, and welfare. Primary education is free and compulsory for those between the ages of 6 and 13. The nation's facilities and teaching staff are inadequate, however, and can accommodate only about three-fourths of the school-age children. Of those who attend school, most leave before completing the secondary level.

Institutions of higher learning include the Nicaraguan branch of the Central American University, in Managua, founded in 1961, and the National Autonomous University of Nicaragua, at León, founded in 1812.

After decades of neglect by the Somoza administrations, health services were at first improved under the Sandinista government. Welfare services are offered by both the labour unions and the government, and a national social security program has been available to all Nicaraguans since 1982. By the late 1980s, turmoil caused by war and economic failures had produced a reversal of many of the social gains.

Cultural life. Nicaragua's long periods of unstable government and dictatorial rule, during which censorship was often practiced, hampered the growth and development of cultural and intellectual expression in the country. Nevertheless, some notable individual artists and institutions have emerged, and, of course, the country has a long tradition and strong interest in pre-Columbian art, which is the basis of a rich present-day American Indian folk art.

Literature

The arts. Nicaragua's literary tradition was consolidated in the late 19th and early 20th centuries with the poetry and prose of Rubén Darío, known as the "prince of Spanish-American literature." Later writers of importance include Santiago Argüello, Lino Argüello, Salomón de la Selva, and Salvador Buitrago Díaz. José de la Cruz Mena is to Nicaraguan music what Darío is to its literature. Although he has not received international recognition, his waltzes are masterpieces of melody and harmony. The composers Edwin Krüger and José Ramírez excel in folkloric music.

Cultural institutions. The National Library at Managua maintains branches in most of the major cities. The National Museum, with a strong concentration on American Indian culture, is located in the capital city, and the Tenderi Museum of Indian artifacts is at Masaya.

Recreation. Fiestas based on religious holidays occur throughout the year and are of particular significance in rural areas. They often involve gambling games and sporting activities, including the nation's favourite, baseball.

Media. Nicaragua has several daily newspapers, among which are *La Prensa, El Nuevo Diario,* and *Barricada,* the official Sandinista organ. There are two major television stations and numerous radio stations. (R.A./B.Ni.)

For statistical data on the land and people of Nicaragua, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

The Spanish conquistador Gil González Dávila made the first attempt to conquer what is now Nicaragua in 1522. Though he claimed to have converted some 30,000 American Indians, carried off 90,000 pesos of gold, and discovered a possible transisthmian water link, González was eventually run out of Nicaragua by angry native inhabitants. Some of the latter were commanded by Nicarao, from whom the country's name derives. It was not until 1524, under Francisco Hernández de Córdoba, that permanent colonization began.

The Spanish conquest was a disaster for the native population of Nicaragua's Pacific region. Within three decades, an estimated Indian population of one million plummeted to a few tens of thousands, as approximately half of the indigenous people died of contagion with Old World diseases, and most of the rest were sold into slavery in other New World Spanish colonies. Few were killed in outright warfare.

British-Miskito Indian alliance

After the initial depopulation, Nicaragua became a backwater of the Spanish empire. In this setting, two colonial cities, Granada and León, emerged as competing poles of power and prestige. The former derived its income from agriculture and trade with Spain via the San Juan River; the latter came to depend on commerce with the Spanish colonies of the Pacific coast. Both tiny outposts were subjected to frequent pirate attacks. Late in the 17th century, Great Britain formed an alliance with the Miskito Indians of the Caribbean coastal region, where the community of Bluefields had been established. The British settled on the Mosquito Coast, and for a time (1740–86) the region became a British dependency.

Independence. In 1811, inspired by struggles in Mexico and El Salvador, revolutionaries deposed the governing intendant of Nicaragua. León, however, soon returned to the royalist cause, and Granada bore the brunt of the punishment for disobedience. In 1821 León rejected and Granada approved the Guatemalan declaration of independence from Spain. Both accepted union with Mexico (1822–23), but they fought one another until 1826, when Nicaragua took up its role in the United Provinces of Central America. After Nicaragua seceded from the federation in 1838, the rivalry between León, identified with the Liberal Party, and Granada, the centre of the Conservative Party, continued.

Foreign intervention. After the withdrawal of Spain, relations between the "king" of the Mosquito Coast and the British government strengthened until again there were British officials in Bluefields. In 1848 they seized the small Caribbean port of San Juan del Norte, renaming it Greytown. The discovery of gold in California drew attention to the strategic position of Nicaragua for interoceanic traffic, and Cornelius Vanderbilt's Accessory Transit Company began a steamship and carriage operation between Greytown and the Pacific. In 1856 William Walker, an American who had been invited to assist the Liberals in warfare (1855), made himself president of the country, but he was routed a year later by the efforts of the five Central American republics and the transit company.

Conservatives ruled from 1857 until 1893, bringing relative peace but little democracy to Nicaragua. As a compromise between Granada and León, Managua was made the capital in 1857. In 1860 a treaty with Great Britain provided for the nominal reincorporation of the east coast with the rest of the nation but as an autonomous reservation. Complete jurisdiction over the Miskito people was not established until the Liberal presidency (1893–1909) of José Santos Zelaya.

Zelaya, though a dictator, was a committed nationalist. He promoted schemes for Central American reunification and refused to grant the United States transisthmian canal-building rights on concessionary terms, thus encouraging the United States to choose Panama for the project. This, plus rumors that Zelaya planned to invite Japan to construct a canal that would have competed with the U.S. waterway, caused the United States to encourage Zelaya's Conservative opposition to stage a revolt. When two U.S. citizens who participated in the revolt were executed, the United States landed marines in Bluefields and thus blocked a Liberal victory. Though Zelaya resigned, the United States refused to recognize his successor, José Madriz (1909–10). Further civil war led to the presidency of a Conservative, Adolfo Díaz (1911–17), for whom the U.S. Marines intervened in 1912. A 100-man guard at the U.S. embassy symbolized that country's support also for Conservative presidents Emiliano Chamorro (1917–21) and his uncle Diego Manuel Chamorro (1921–23). The Bryan-Chamorro Treaty, signed in 1914 and ratified in 1916, gave the United States exclusive canal privileges in Nicaragua (to prevent a competing canal from being built) and the right to establish naval bases.

The U.S. Marine guard's withdrawal in 1925 led quickly to another crisis, with Emiliano Chamorro in rebellion against a new regime. Díaz returned as a compromise president (1926–28), reinforced in 1927 by 2,000 U.S. Marines. The Liberal leaders Juan Bautista Sacasa, José María Moncada, and Augusto César Sandino rose in rebellion, but after six months Sacasa and Moncada made peace, and subsequent elections under U.S. auspices brought the presidency to both of them (Moncada, 1928–33, and Sacasa, 1933–36). Sandino, however, fought on as long as the Marines remained in the country.

The Somoza years. The Marines withdrew upon the inauguration of Sacasa, and Sandino submitted to his government. A Nicaraguan National Guard (Guardia Nacional), trained by the U.S. Marines and commanded by General Anastasio Somoza García, was now responsible for maintaining order in the country. In 1934 high-ranking officers led by Somoza met and agreed to the assassination of Sandino. Somoza then deposed Sacasa with the support of factions of both Liberals and Conservatives, and in a rigged election, he became president on Jan. 1, 1937.

Somoza (known as Tacho) revised the constitution to facilitate the consolidation of power into his own hands and ruled the country for the next two decades, either as president or as the power behind puppet presidents. Export activities grew from the 1930s onward. However, the Somoza family and their associates, rather than the Nicaraguan people was a whole, were the main beneficiaries of the country's income.

Assassination of Somoza

On Sept. 21, 1956, a day after Somoza's Nationalist Liberal Party of Nicaragua (Partido Liberal Nacionalista de Nicaragua; PLN) had nominated him for another term, a liberal poet named Rigoberto López Pérez shot the president, who died eight days later. Congress at once gave Luis Somoza Debayle his father's position, and in February 1957 he was elected dubiously to his own term (1957–63). Somoza Debayle ruled more gently than his father had. He accepted a settlement in favour of Honduras of a long-standing border dispute between the two countries (1960) and cooperated with the United States in the so-called Bay of Pigs invasion of Cuba (1961). In 1961 three Marxists, including Carlos Fonseca Amador, founded the guerrilla Sandinista National Liberation Front (Frente Sandinista de Liberación Nacional; FSLN) in opposition to the regime. Named for Augusto César Sandino, its members are called Sandinistas.

A proposal that the 1963 election be supervised by neutral observers was refused by Somoza Debayle, and the two parties in consequence declined to nominate candidates. René Schick Gutiérrez, the PLN candidate and a friend of the Somozas, thus easily won the presidency in a virtually uncontested election. The economy was prospering, though the poverty of the masses remained relatively unchanged. Lorenzo Guerrero Gutiérrez became president for one year upon the death of Schick in office. Luis Somoza Debayle died in 1967, and in that year his younger brother, Anastasio Somoza Debayle (called Tachito), in another rigged election, defeated Fernando Agüero Rocha for the next term. In 1970 the Bryan-Chamorro Treaty was abrogated.

On May 1, 1972, constitutionally ineligible to succeed himself, Somoza relinquished the presidency to a triumvirate (composed of Agüero and two leaders of Somoza's own party). On December 23 an earthquake in the city of Managua left 6,000 persons dead and 300,000 homeless. Somoza (commanding the National Guard) took charge as the head of a National Emergency Committee. Agüero, who protested, found himself replaced (March 1, 1973) on the triumvirate. The population suffered from the destruction as Somoza and his friends profited privately from international aid programs. In March 1974 a new constitution (the country's 10th) made it possible for Somoza to be reelected president, and on December 1 he again assumed the office.

Before the end of the year, two genuine opposition groups attracted wide attention—the Sandinistas and the organization founded by Pedro Joaquín Chamorro, editor and publisher of *La Prensa* of Managua, called the Democratic Union of Liberation (Unión Democrática de Liberación; UDEL). In December 1974 the Sandinistas staged a successful kidnapping of Somoza elites, for which ransom and the release of political prisoners was obtained. In response, the regime embarked on a two-and-a-half-year counterinsurgency effort that, in addition to leading to the death of Carlos Fonseca in 1976, took the lives of thousands of peasant noncombatants. In 1977 a group called Los Doce (The Twelve) sought an anti-Somoza alliance to include UDEL, the Sandinistas, and other organizations. Assassins murdered Pedro Joaquín Chamorro on Jan. 10, 1978, and a general strike and violence followed. On August 22 the Sandinistas occupied the national palace, holding more than 1,000 hostages for two days and winning most of their demands. Although the National Guard regained partial control, the insurrection spread, with another general strike and the Sandinistas seizing and holding several major cities. The uprising was eventually quashed, at the cost of several thousand lives. The following June the FSLN staged its final offensive. City after city fell to the insurgents, backed by tens of thousands of local civilian combatants. On July 17 Somoza resigned and fled the country; two days later, the Sandinistas entered Managua and accepted the surrender of what was left of his army, ending the long years of Somoza rule. (F.D.P./T.W.Wa.)

End of the Somoza regime

The Sandinista regime. The new government inherited a devastated country. About 500,000 people were homeless, more than 30,000 had been killed, and the economy was in ruins. In July 1979 the Sandinistas appointed a 5-member Government Junta of National Reconstruction. The following May it named a 47-member Council of State, which was to act as an interim national assembly. In 1981 the junta was reduced to 3 members and the council increased to 51.

In 1979–80 the government expropriated the property held by Anastasio Somoza Debayle, members of his government, and their supporters. Local banks and insurance companies and mineral and forest resources were nationalized, and the import and export of foodstuffs were placed under government control. The Statutes on Rights and Guarantees, which acted as the country's new constitution, assured basic individual rights and freedoms. The government disclaimed any responsibility for the assassination of Somoza, on Sept. 17, 1980, in Asunción, Paraguay.

Although the Sandinista government expanded ties with noncommunist nations, it also established close relations with Cuba and other socialist-bloc nations. The U.S. government interpreted this posture as an indication of further communist expansion in the Western Hemisphere. In 1981 the United States suspended economic aid to Nicaragua, and later that year U.S. President Ronald Reagan authorized nearly $20 million for the creation and arming of an irregular force of Nicaraguan counterrevolutionaries to fight the Sandinistas. These insurgents, who came to be called Contras, established bases in the border areas of Honduras and Costa Rica. The Contra army was expanded to around 15,000 soldiers by the mid-1980s. Eventually, the Nicaraguan government expanded its military forces, acquired crucial equipment such as assault helicopters, and implemented counterinsurgency strategy and tactics, which enabled it in the late 1980s to contain and demoralize the Contras. In 1987, the U.S. Congress voted against supplying further military aid to the Contras. That August, the first of several Central American peace agreements was signed, which gradually moved the locus of Nicaraguan conflict from the military to the political sector.

Daniel Ortega Saavedra

Meanwhile, on Nov. 4, 1984, the FSLN and its presidential candidate, Daniel Ortega Saavedra, won 63 percent of the vote in an election that international observer teams had deemed fair and competitive. Ortega was inaugurated in January 1985, and the new Constituent Assembly produced a constitution two years later, which called for regularly held elections, the first for national office to take place in 1990.

The Reagan administration, which had denounced the 1984 election as a sham, continued to exert pressure on the Sandinista government. A U.S. trade embargo was declared in 1985. Washington used its leverage within the World Bank and the Inter-American Development Bank to block most Nicaraguan loan requests from 1982 onward. These measures and the damage and economic dislocation caused by the civil war combined with Sandinista economic errors to cause Nicaragua's economy to plummet from 1985 onward. An annual inflation rate of more than 30,000 percent in 1988 was followed by harsh and unpopular austerity measures in 1989. Government programs in health, education, housing, and nutrition were drastically curtailed.

Against this background, and under intense international observation, the 1990 general elections were held. Contra activity increased during the electoral period. On Feb. 25, 1990, the U.S.-endorsed and -financed National Opposition Union (Unión Nacional Oposición; UNO) coalition and its presidential candidate, Violeta Barrios de Chamorro, the widow of the martyred newspaper editor, won an upset victory, and a peaceful transfer of administrations took place. A cease-fire with the Contras was implemented, and Chamorro eventually was able to reach an agreement with them to lay down their arms, although some factions were considerably reluctant to do so. (T.W.Wa.)

For later developments in the history of Nicaragua, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

Panama

The Republic of Panama (Spanish: República de Panamá) is a country in Central America that is situated on the S-shaped Isthmus of Panama that joins North and South America. Its area is 29,762 square miles (77,082 square kilometres), of which 29,187 square miles is on the mainland and 575 square miles is divided among islands. The Panama Canal, which connects the Atlantic and Pacific oceans, runs through the country.

Location and boundaries

Panama extends for about 480 miles (770 kilometres) from the Costa Rican frontier in the west to the Colombian border in the east. The shortest distance across the isthmus is 30 miles, stretching from the mouth of the Nergalá (Necategua) River, which flows into the Gulf of San Blas on the northern (Caribbean) shore, to the mouth of the Chepo River on the Pacific coast. Panama is bounded on the north by the Caribbean Sea, on the south by the Pacific Ocean, on the east by Colombia, and on the west by Costa Rica. Panama City, the capital, is located on the Pacific coast to the east of the Panama Canal.

About one-third of the economically active population of Panama is engaged in agriculture and cattle raising. Panama also has a strong services sector, especially in transport, banking, and international trade, which has developed on the basis of the country's strategic location.

PHYSICAL AND HUMAN GEOGRAPHY

The land. *Relief.* A central spine of mountain ranges extends almost the entire length of Panama, dividing the country into Atlantic- and Pacific-facing slopes. The two principal mountain ranges, the Tabasará Mountains (Cordillera Central) in the west and the Cordillera de San Blas in the east, are separated near the centre of the country by a saddle of lower land. This depression (the Panama Canal site) divides the country again—roughly into western and eastern halves. Of the four quadrants thus formed, the southwestern is the most important: with the canal area itself it accounts for most of Panama's population and commerce. The country's highest peak is an inactive volcano, Barú (Chiriquí; 11,401 feet [3,475 metres]).

Paralleling the principal mountain chains, a lower mountain arc extends along Panama's southern coast. It appears only in well-separated segments—for example, on Azuero Peninsula as the Canajagua Massif and, in eastern Panama, as the Sierra de Jungurudó, Sapo Mountains, and the Majé Mountains.

The land surface may also be divided roughly into two geologic categories: the above-named highlands and mountains, which are made up mainly of igneous rocks (predominantly old volcanics), and lowlands, which commonly consist of sedimentary rocks, such as slates and shales. Such lowlands include the plains of Panamá and Chiriquí provinces, the plains and hills of Colón province, the Chepo and Chucunaque river basins in the east, and the narrow northeastern plains of the Caribbean region.

The irregular Pacific coastline (extended by the Azuero Peninsula and the Gulf of Panama and by numerous headlands and bays) is considerably longer than the Caribbean coastline; the largest embayment on the Caribbean side is Chiriquí Lagoon. Similarly, the continental shelf is much wider on the Pacific side than on the Atlantic.

Offshore islands

Most of the republic's more than 1,600 islands are off its Pacific coast; they include the Perlas Archipelago (Pearl Islands) and the islands of Toboga, Cêbaco, Parida, Jicarón, and Coiba, the largest. The principal archipelagoes off the Caribbean coast are those of Bocas del Toro and San Blas.

Drainage and soils. Of Panama's many rivers, those that flow to the Caribbean include the Sixaola, Changuinola, Indio, Cricamola, La Miel, and Chagres. Rivers flowing to the Pacific include the Chiriquí Viejo, Santa María, Chepo, and Chucunaque. A few are navigable during the rainy season; the Tuira is navigable for about 40 miles and the Chepo for about 20 miles.

Soils are commonly reddish to brown and rich in clay. They vary in fertility, and in many areas crops can be grown continuously only if fertilizers are applied. On poorer soils, a shifting subsistence agriculture is practiced. Under this system small plots are cleared, cropped for a few years, then abandoned until their natural fertility is restored—a practice called *roza* in Panama.

Areas of alluvial soils (which develop from clay, silt, sand, and gravel deposited by streams) are especially fertile. Such soils, however, are not extensive, being limited to the lower parts of river valleys. The commercial banana plantations around Puerto Armuelles and in western Bocas del Toro province are mainly on alluvial soils. Some of the soils along the inland edges of coastal mangrove swamps have also proved quite productive. Soils that develop from volcanic ash are exceptionally fertile.

Climate. The climate is markedly different on the Atlantic and Pacific sides of western Panama, especially in the amount and seasonal distribution of rainfall. On the Caribbean slopes of the Tabasará Mountains, which face the rain-bearing trade winds, average rainfall is approximately twice as heavy as on the leeward Pacific slopes. The Caribbean coast receives between 60 and 140 inches (1,500 and 3,500 millimetres) per year, while the more populated Pacific region receives between 45 and 90 inches (1,140 and 2,290 millimetres). Furthermore, on the Caribbean side, rain falls almost throughout the year, while on the Pacific side—in Chiriquí province, for example—there is a pronounced dry season lasting from January to April; March is usually the driest month. In accordance with this rainfall pattern, tropical rain forest is the typical vegetation on Caribbean slopes, while southward in the seasonally dry areas found between the Tabasará Mountains and the Pacific shoreline, savannas are typical. Toward the Colombian border both sides of the isthmus have year-round rainfall and extensive rain forests.

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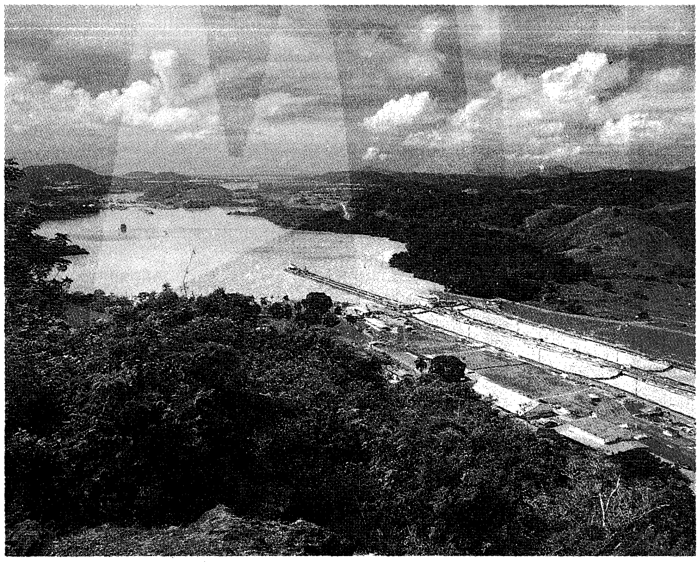

The Pedro Miguel Locks raises or lowers ships on the Pacific side of the Panama Canal.

Temperatures

Rarely on either coast does the mean temperature of the coldest month drop below 78° F (26° C). On mountain slopes three altitudinal climatic zones are recognized in Panama: a low, hot zone, lying at altitudes below 2,300 feet (700 metres), which makes up more than 85 percent of the country's territory; a temperate zone, at altitudes of between 2,300 feet and 4,900 feet, which makes up about 10 percent of the land; and a cold zone, lying above 4,900 feet and covering less than 5 percent of the total area. Each zone has its characteristic plant growth and crops; thus, cacao and bananas are produced mainly in the hot lowlands, while coffee is grown in the temperate zone.

Plant and animal life. Despite its relatively small area Panama has a great variety of landscapes and habitats, including tropical rain forest, savanna (tropical grassland), cool montane forest, tidal lands covered by stilted mangrove trees, coral reefs, and beaches.

Because of its ancient role as a land bridge over which species have migrated between the continents, the isthmus is home to a rich intermixture of plant and animal life. As a result of this interchange, Panama's wildlife includes such mammals as sloths, anteaters, and armadillos—which derive from South America—as well as the jaguar, tapir, and deer—which came originally from North America. Also among Panama's rich fauna are several species of giant sea turtle, which lay their eggs on the beaches. Few comparable areas have a larger number of bird species.

Settlement patterns. Almost half of the Panamanian population is rural, with people living in hamlets or isolated dwellings and engaged in subsistence agriculture. The densest rural population is found in the broad savanna area of plains and low hills lying between the Azuero Peninsula and the Tabasará Mountains. The least densely settled areas are in the eastern third of the country, especially in the large province of Darién, and on the Atlantic slopes of the Tabasará Mountains.

The percentage of the population living in cities has grown rapidly in recent decades, and in the early 1980s the country became predominantly urban. The most densely settled part of Panama is a broad coast-to-coast strip of land that crosses the country where it is at its lowest and narrowest: in the vicinity of the canal. This general area, partly drained by the Chagres River, is known as the Chagres, or Route, region. It includes the cities of Panama City and Colón, the district of San Miguelito, and the towns of Balboa, La Chorrera, Gamboa, and Cristóbal. Panama City, situated on the Pacific coast overlooking Panama Bay, is the dominant population centre and the centre of industrial, commercial, political, and cultural activities. San Miguelito is the second largest urban centre and Colón the third largest.

The Canal Zone

Panama Canal. The Panama Canal, built by the United States and operated continuously since its opening in 1914, has made Panama an international shipping and trade centre. From 1903 until 1979 a 10-mile-wide strip of land lying on either side of the canal, the Canal Zone, was controlled by the United States. By treaties signed between the two countries in 1977, the Canal Zone was abolished in 1979, and Panama reasserted its sovereignty over the Canal Zone, with Panama and the United States to share in the operation and defense of the canal itself through 1999.

Since 1979 the canal has been operated by the Panama Canal Commission, a joint U.S.-Panamanian agency. There are several thousand U.S. civilians and military personnel living in the area who help operate and protect the canal, and the United States, by agreement with Panama, has continued to maintain military bases there.

The people. *Ethnic and linguistic groups.* In the 16th century, when the Spaniards first went to the isthmus, it was occupied by Cuna, Guaymí, Chocó, and other Indian groups. Mestizos resulted from the miscegenation that occurred between the Spanish and the Indians. During colonial times people from Africa were brought as slaves, and still other mixed racial types developed as the Africans entered the society. During the 19th century, with the construction of the Panama City-Colón railroad, new groups arrived—North Americans (primarily from the United States), French, and Chinese. During the construction of the canal, more U.S. nationals went to Panama, as did blacks (from British Barbados and Jamaica), Spaniards, Italians, and Greeks.

Though they are but a small fraction of the total population, Indians are found scattered over a considerable area, often in rain forests on rough terrain. The most numerous of the Indian groups are the Guaymí, who live in the western provinces of Chiriquí, Bocas del Toro, and Veraguas. Next in numbers are the Cuna, who live mainly in the San Blas Archipelago and on the coast nearby. The Chocó live mainly in the province of Darién. All these groups preserve their native languages, but many persons in each can also speak Spanish. Although most are engaged largely in subsistence agriculture, fishing, and hunting, a number of Cuna are traders, sailors, and mechanics and work in other occupations. Also, many Guaymí work on the banana plantations of western Panama.

Mestizos are the largest population group in Panama. They predominate in the savannas to the west of the canal and in the central provinces of Panamá and Colón, where they have intermarried with people of African ancestry.

People of African descent who were brought to the isthmus during the period of Spanish colonization are settled throughout much of Panama—for example, in the hot lowlands of the Chagres River basin, in the eastern province of Darién, and in the Caribbean lowlands. Black peoples from the British West Indies are more recent arrivals and make up only a small minority. They live primarily in Panama City and in Bocas del Toro province.

U.S. influence

The United States has influenced both the economy and the culture of Panama. Americans live primarily in the canal area and in Panama City. Other significant minorities in Panama are Chinese, East Indians, Jews, and Middle Easterners, all of whom play an important role in commerce and industry and who participate in the country's political and professional life.

Spanish is the official language of Panama and is spoken by almost 90 percent of the people. English is spoken mainly by West Indians, who make up about 8 percent of the population; about 5 percent speak Indian languages.

Religion. Roman Catholicism is the religion of most Panamanians, but Protestant enclaves have appeared, for example, within the black population of Bocas del Toro province and among Indian groups. Panama's constitution guarantees freedom of worship.

Demographic trends. A striking attribute of Panamanian demography is the high rate of population increase. The population grew from 336,000 persons in 1911 to almost 2,500,000 by the end of the century. An important factor in population growth has been immigration, which was substantial during World War II but has since decreased. Two other demographic changes are noteworthy: the movement of rural people within the country in search of land and the movement of young people to urban areas—especially to Panama City and Colón—in search of better living and economic conditions.

The economy. Most of the work force is engaged in agriculture (including cattle raising, fishing, and forestry) and in services, manufacturing, and trade.

Resources. Clay, limestone, and salt are the leading mineral products, and gold, ferrous sand, and manganese have been produced on a small scale. There are commercially significant, but largely unexploited, deposits of copper, including the Cerro Colorado find in Chiriquí province. Deposits of bauxite, phosphates, and coal have been minimally exploited, as have various construction materials such as stone and gravel. Oil has been located offshore, in both the Pacific and the Caribbean.

Energy. Electricity is distributed by the Institute of Hydraulic Resources and Electrification. Much of Panama's electricity is generated by hydroelectric plants. The first plants were opened in 1975 at La Yeguada in Veraguas province and in 1976 on the Chepo River; the largest, at La Fortuna, opened in 1984.

Agriculture, forestry, and fishing. A seminomadic and subsistence agriculture exists alongside large commercial enterprises that use modern agricultural methods. Land tenure, allowing small farmers to acquire land, was instituted in 1962, but by late in the 20th century less than one-third of all farms were owned by the people working them. The most common agricultural products include rice, corn (maize), sugarcane, beans, bananas, and plantains. The growing conditions of the isthmus favour the cultivation of tropical fruit, although commercial production for most of them is limited. The commercial cultivation of bananas, coffee, cacao, sugarcane, and other crops has increased considerably.

Principal crops

Livestock raising (cattle, pigs, and poultry) is an important and long-established economic activity. The centres of the cattle industry, and the country's largest cattle-raising farms, are in the southwestern savannas—in Chiriquí, Los Santos, and Veraguas provinces. The quality of meat produced has been greatly improved by the introduction of better breeding stock (including Brahman cattle from Asia) and nutritious forage grasses from Africa. Beef exports began in the 1970s. Hides, however, are a much more economically significant by-product of cattle raising.

The raising of pigs has increased because of a growing urban market and relatively high prices. The provinces of Panamá, Los Santos, Chiriquí, and Veraguas raise the most pigs; Panamá, Coclé, and Colón are the largest centres for the raising of chickens and the production of eggs.

Considering Panama's abundant and valuable forest resources, there has been little use of the nation's native trees for industrial purposes. Mahogany, tropical cedar, cativo (a large tree belonging to the legume family), and other woods supply domestic sawmills, and some logs are exported. The government has found it difficult to supervise uses of the nation's forests; for instance, large tracts of Panama's fine rain forests are being felled and burned and are being replaced by pastures for an expanding cattle industry.

Fishing has developed rapidly as a commercial venture. Shrimps, prawns, and lobsters are among Panama's most important exports, with several thousand tons caught yearly. The shrimp and fish (mainly anchovies and herring) catch is taken mostly from the shallow waters of the continental shelf in the Gulf of Panama. Lobsters are caught mainly off Bocas del Toro and San Blas on the Atlantic side. Aquaculture has been successfully developed in Panama.

Leading manufactured products

Industry. Although there has long been a considerable international trade based on the country's key location as a crossroads, modern industrial development began mainly after World War II. The leading industries are cement manufacture (in the limestone areas near the Trans-Isthmian Highway), fish processing, sugar refining, processing of general food products, banana and cocoa production, oil refining (near Colón), and the production of gas and electricity. Other important industries include clothing, shoe and leather goods manufacturing, brewing, distilling, and the production of furniture, paper, hides, chemicals, and cigarettes.

The main manufacturing centre is Panama City, and there is some industry at Colón and David. Tourism plays a major role in the country's economy, providing funds for industrial development.

Finance. Government laws have promoted the development of offshore banking. Financial services are provided by representative branches of several major Latin American, North American, and European banks. Some of these banks have branches in the provinces and provide loans for industrial, agricultural, and cattle-raising ventures.

Trade. The main products purchased from abroad are crude oil, fuels and lubricants, machinery, transportation equipment, manufactured goods, food products, and chemicals. Most of the imports are from the United States, the Colón Free Zone (see below), Mexico, and Japan. Exports include food products, such as bananas, shrimps, and coffee, as well as cattle hides, clothing, and pharmaceuticals. The United States takes a large part of these exports; other notable buyers of Panamanian products are Germany and Costa Rica.

The Colón Free Zone, established in the mid-20th century at the northern end of the canal, has become increasingly important as a manufacturing, warehousing, and reexport centre. Its several hundred plants produce chemical products, manufactured articles, machinery, and transportation equipment. Raw materials or semifinished products come from the United States, Japan, the United Kingdom, and Germany. Finished products are sent largely to the countries of Central and South America, but some are used domestically.

Transportation. Historically, Panama has been indeed one of the world's great passageways. In ancient times the isthmus was an intercontinental thoroughfare for migrating plants, animals, and peoples. Today, Panama links not only the continents but the oceans as well. In addition to the canal, Panama is the site of an interoceanic highway, railroad, and oil pipeline.

The country has several thousand miles of roads, one-third of which are paved. The main routes extend from the capital to the central provinces (Coclé, Herrera, Los Santos, and Veraguas) and to the western section of the country (Veraguas and Chiriquí provinces). The principal highways include the Trans-Isthmian Highway, running across the isthmus from Panama City to Colón, and the Inter-American Highway, the Panamanian section of the Pan-American Highway; the latter has been completed from the Costa Rican border eastward, through Panama City, to a point near Yavisa in Darién province.

Inter-oceanic pipeline

In 1982 an interoceanic oil pipeline was built in western Panama, connecting the Pacific and the Atlantic coasts. The line extends some 80 miles from Puerto Armuelles over the Tabasará Mountains to Chiriquí Lagoon. A road was built through the mountains paralleling the pipeline.

A 49-mile transcontinental railroad, the Panama-Colón line, was completed in 1855 by a U.S. company. It is now administered by the government of Panama. The Chiriquí National Railroad operates both passenger and freight service between La Concepción and Puerto Armuelles and freight service between David and Pedregal on the Pacific. In Bocas del Toro province there is a railroad serving the banana-growing area.

The Panamanian coastline has many natural harbours that have excellent conditions for sheltering vessels, but the best ports are those at either end of the Panama Canal—Cristóbal, near Colón, and Balboa, near Panama City. Other important ports, primarily serving the banana trade, include Almirante and Puerto Armuelles. Coastal shipping is important for Panamanian agriculture.

There are more than 100 airports in Panama and both domestic and international Panamanian airlines. Tocumen International Airport (officially General Omar Torrijos Herrera International Airport), is Panama's major airfield; it is located 16 miles from the capital and is served by several international airlines. There are regular domestic flights from Panama City to various parts of the country.

Administration and social conditions. *Government.* After a coup d'état by the National Guard in 1968, the legislative branch of government was suspended, and Panama was administered by a provisional government led by General Omar Torrijos. A new constitution in 1972, the fourth in Panama's history, gave Torrijos virtually complete control over the government but also established an elected National Assembly of Municipal Representatives. The constitution was amended in 1978 to provide for a gradual return to democratic government within six years. Further constitutional amendments were approved in 1983, but constitutional government did not return to Panama until after the downfall of Torrijos' successor, General Manuel Antonio Noriega Moreno, in 1990.

Under the constitutional revision of 1983, executive power is exercised by the president and two vice presidents, elected by popular vote for a five-year term. The president appoints the Cabinet. The constitution established a unicameral Legislative Assembly consisting of 67 members elected for five-year terms. The assembly has powers to initiate legislation, rule on international treaties, approve the budget, and establish political divisions.

Local government

The country is divided into nine *provincias* and the *comarca* (sector) of San Blas. The *provincias* are divided into *distritos municipales* (municipal districts), which are subdivided into *corregimientos* (magistracies). The head of each *provincia* is the governor, appointed by the president. San Blas is governed by tribal leaders (caciques) of the Cuna Indians.

Justice. Judicial power rests with the Supreme Court of nine members. They are appointed for 10-year terms by the president with the approval of the Legislative Assembly. The Supreme Court is composed of separate divisions for civil, penal, and administrative cases. The justice system also includes several types of lower courts. Police services are operated by the National Guard.

The armed forces. The Panamanian Defense Force was created in 1983, incorporating the National Guard, the Air Force, the Navy, the Panama Canal Defense Force, the police, the National Department of Investigations, and the Immigration Department. Historically, the military has frequently used its power to take over and run the country.

Education. All children have the right to receive an education from the state without discrimination by race, sex, or economic standing. Elementary education is compulsory and free from ages 6 to 15. Illiteracy has been decreasing. Private schools cannot teach in a foreign language without the consent of the Ministry of Education.

The institutions of higher education include the state-run University of Panamá (founded 1935) and the privately operated University of Santa María la Antigua (1965), both in Panama City. The University of Panamá, the official university, also has branches in several provinces. A polytechnic school was founded in Panama City in 1981.

Health and welfare. The government has built a system of hospitals and health centres and has granted funds for their improvement. Hospitals and hospital-clinics are supplemented by regional health centres and by mobile medical units. A social security system provides a wide range of benefits; most permanent employees are covered. The social security administration maintains hospitals and clinics in Panama City and Colón for its beneficiaries and also supports medical centres and clinics in several places in the interior. The St. Thomas Hospital, in Panama City, provides treatment for those unable to pay hospital expenses. There are some private hospitals.

About 90 percent of the working population is protected by a minimum wage law; the government also regulates prices. Both salaries and wages are subject to great fluctuations, however. Most workers and professionals are union members, and collective labour contracts are usual. The government also has established the payment of an extra month's salary, called *décimo tercer mes* ("13th month"), each year for all workers.

Cultural life. Distinctive Panamanian cultures can be divided regionally. The cosmopolitan urban culture near the canal contrasts with the rural culture of the southwestern savannas. The latter area, with its cattle ranches and horsemanship, is something of a centre of Hispanic tradition. Old folk songs and handicrafts (like the straw-hat industry) are preserved there—for example, around the towns of Chitré and Las Tablas. Also culturally distinctive are the territories of the various Indian groups, each with its language and handicrafts, such as the needlework clothing designs (*molas*) sewn by Cuna women and the netted carrying bags made by the Guaymí. Other areas of cultural interest include the Caribbean islands of Almirante Bay with their Antillean customs.

Rural culture

The arts. To a large extent national artistic achievements did not begin to develop in Panama until after independence in 1903. Before that time artistic themes were mostly European or church-related. Some progress has been made in national expression since that time, and there has been some international recognition of Panamanian artists. The local culture is, however, supplemented by visits to the population centres from musical and theatrical groups and by poets, sculptors, and other artists.

Cultural institutions. Among the notable cultural institutions are the Panamanian Art Institute (Panarte), the National Institute of Music, the National School of Music, the National School of Plastic Art, the National School of Dance, and the National School of Theatre. Also of note are the National Commission on Archaeology and Historic Monuments, the National Museum of Panama, and the Panamanian Academy of History.

Recreation. The Panamanian Institute for Tourism encourages the preservation of traditional holidays, folk music, and folk dances. Panamanian carnivals are known for their beautiful costumes and joyous music. The National Institute of Culture promotes many types of events, including concerts, theatre, and art expositions. The National Institute of Sports promotes intraprovincial sporting events as well as Olympic competition. The most popular spectator sports in Panama, in contrast to those of other Central American countries, are baseball, boxing, horse racing, and basketball, rather than soccer. There are numerous attractive beaches and sportfishing waters, as well as cool vacationlands on the slopes of Barú Volcano. During the 1970s and '80s about one-sixth of the area of the country was designated as national parks or preserves.

Press and broadcasting. The principal newspapers are published in Panama City. Panama has many radio stations, half of them located in Panama City. There are three television networks, with several stations. The United States also operates a television station from Fort Clayton. (N.M.C./B.L.G.)

For statistical data on the land and people of Panama, see the *Britannica World Data* section in the BRITANNICA WORLD DATA ANNUAL.

HISTORY

Exploration, conquest, and settlement. The Atlantic coast of the Isthmus of Panama was first explored by Rodrigo de Bastidas in 1501, in the company of Juan de la Cosa and Vasco Núñez de Balboa. In 1510 Diego de Nicuesa, another Spanish explorer, established the settlement of Nombre de Dios at the mouth of the Chagres River, and to the southwest Alonso de Ojeda that of San Sebastian de Urabá. Because of the presence of bellicose Indians, the colony was moved, at the instigation of Balboa, northeast across the Atrato River and was given the name Santa María la Antigua del Darién. It became the first permanent settlement on the isthmus and the focus of jealous intrigues centring around Balboa.

Early explorers

As head of the colony, Balboa, by the use of persuasion and force, brought most of the Indians under submission. Some of them revealed to him the existence of a large sea and of a gold-rich empire to the south, perhaps that of the Inca. On Sept. 26, 1513, Balboa reached the sea and claimed the Pacific Ocean for his king. Returning to Santa María, he encountered much resistance from the Indians. Soon afterward Balboa was executed for insurrection on orders of the new governor, Pedrarias Dávila, known as "the Cruel," who distrusted Balboa and feared his rivalry.

In 1519 the population of Santa María moved to the new town of Panama (the first European settlement on the west coast of the hemisphere), which became the centre of commercial activity and the springboard for the conquest of Peru. The colony, also called Panama, became an important part of Spain's mercantile system, attaining the rank of audiencia in 1538, while Nombre de Dios, now resettled and linked to Panama by road, was renowned for its ferias (exchange markets). With the final destruction of Nombre de Dios in the late 16th century by Francis Drake, the English buccaneer, commercial activity was moved to the hamlet of Portobelo, overlooking the calm bay discovered by Christopher Columbus in 1502. Portobelo then became the centre of Spain's commerce in the New World and the site of great ferias.

Panama and Portobelo continued to attract the attention of English raiders, however, and both were visited with disastrous consequences, including destruction of the former by Henry Morgan (1671) and the latter by Admiral Edward Vernon (1739). In the year of Vernon's raid the colony was reduced in status when Spain made it a part of the Viceroyalty of New Granada. Portobelo was rebuilt in 1751, but by then the Spanish galleons had begun to use the route around Cape Horn, accelerating the city's decline through loss of trade. In 1673 the town of Panama was rebuilt a few miles west of the old town.

Secession from Spain and union with Gran Colombia. As the agitation for independence grew elsewhere in Spanish America, Panama, thanks to the foreign trade permitted by the crown, was prosperous and did not heed the call for insurrection. The viceroy moved his headquarters from Quito to Panama, and the isthmus sent deputies to the Cádiz Cortes (parliament) in Spain. When the Spanish merchants secured the revocation of the royal decree authorizing foreign trade, Panama changed sides. In the autumn of 1821 the colony seceded from Spain and joined the Gran Colombia union. For a time Panama enjoyed the right to elect its own governor, but in 1843 a new constitution returned that power to Bogotá. Soon afterward Panama became a state within Colombia.

Transcontinental railroad and canal projects. The building of a transcontinental railroad was negotiated in 1845. It led to the Bidlack-Mallarino Treaty of 1846, which granted the United States the right of isthmian intervention to protect the line and free transit across the continent. After many vicissitudes, stemming from both political and health problems, the railroad was completed in 1855. While construction was under way, political turmoil raged in the land. Panama inaugurated and discarded 20 presidents (formerly called governors), while New Granada (Colombia) elected, substituted, or deposed 7 in Bogotá.

The railroad helped the gold rushers destined for California after 1848, but it also encouraged canal planners. Ultimately the rights to build the canal were awarded to the French diplomat Ferdinand de Lesseps (who had been instrumental in building the Suez Canal) and his Universal Interocean Canal Company; construction began in 1880. By 1889 disease, chicanery, and financial scandals in France and Bogotá involving the management of the works ruined the corporation.

By 1892, Philippe Bunau-Varilla, a former chief engineer of the canal works, had formed a new canal company, which acquired the assets of the defunct one. After the United States became interested in the canal project, it passed the Spooner Act of 1902, which authorized the U.S. president to buy the assets of the French company. Following up on this, Bunau-Varilla pressured Colombia to negotiate the Hay-Herrán Treaty of 1903, which allowed for the transfer of the canal strip to the United States. The Colombian senate, however, rejected the treaty.

Oppressed by Colombian requisitions of property and the impressment of men as a result of their civil war and influenced by Bunau-Varilla and U.S. interests, a revolutionary junta proclaimed Panamanian independence on Nov. 3, 1903. The Colombian forces sent to crush the rebellion were left at Colón when railroad cars were unavailable. U.S. naval forces also deterred reinforcements sent from Bogotá by land. Bunau-Varilla was appointed by the secessionist junta minister to Washington, D.C., with full powers to negotiate treaties. On Nov. 18, 1903, he and Secretary of State John Hay signed the Hay–Bunau-Varilla Treaty. By this treaty the United States obtained, in perpetuity, the exclusive use, occupation, and control of the Canal Zone, a strip 10 miles wide (5 miles on each side of the waterway). Formal acquisition of the lands took place on May 4, 1904, and the canal was opened on Aug. 15, 1914.

Hay–Bunau-Varilla Treaty

The Republic of Panama. The new constitution authorized the United States to intervene militarily in Panama in order to quell disturbances. It also provided for a centralized government headed by a president who had the authority to appoint and dismiss provincial governors. Manuel Amador Guerrero became the first president. Universal suffrage was adopted in June 1907. Political and economic unrest brought bloodless military interventions by the United States in 1908, 1912, 1918, and 1925.

In January 1931 the government of Florencio Harmodio Arosemena was overthrown in a bloody coup led by Arnulfo Arias Madrid, an event that ushered in an age of political and economic turmoil. The United States acquiesced in this deed and promptly recognized as president the minister to Washington, Ricardo Alfaro, who presided over the orderly elections of 1932, when Harmodio Arias Madrid (brother of Arnulfo) was the winner.

Arias persuaded the United States to relinquish the right of intervention and its priority in seizing lands for canal purposes—resulting in the modification in 1936 of the Hay–Bunau-Varilla Treaty by the Hull-Alfaro Treaty. The United States also increased the annuity paid for the use of the Canal Zone and agreed to build a transisthmian highway. But Arias planted the seeds of future military rule by building an impregnable fortress next to the city jail and by encouraging the militarization of the police.

U.S. aquisition of defense sites

In June 1940 Arnulfo Arias was elected president, and he changed the constitution to extend the length of his term. Before the outbreak of World War II the United States requested defense sites outside the Canal Zone for landing fields, roads, antiaircraft batteries, and warning stations. Arias, who openly sympathized with fascism, demanded compensation in the form of cash and the transfer to Panama of various properties. While in Havana on a private visit, he was removed from office by the military in October 1941, and Ricardo Adolfo de la Guardia became president. After the attack on Pearl Harbor on Dec. 7, 1941, Panama transferred the defense sites to the United States.

During and after the war the United States returned 98 defense sites to Panama but continued negotiations on the remaining 36. Agreement was finally reached, but the Panamanian National Assembly, influenced by the threat of mob violence, rejected the proposals. The United States abandoned the sites in 1947.

In 1945 the National Assembly put an end to the regime of de la Guardia and selected the ambassador to Washington, Enrique A. Jiménez, as president. In the presidential elections of 1948 Domingo Díaz Arosemena was declared the winner by the National Jury of Elections, depriving Arnulfo Arias of a victory. Díaz Arosemena died in office in 1949, and under a mountain of popular protests his two vice presidents first accepted, then refused, office. This action cleared the way for a bizarre maneuver by the election jury, which declared that, after a recount, Arias had won the presidency in 1948.

In 1951 former president de la Guardia returned from exile and began to challenge the government. In May a run on, and the subsequent closing of, the Federal Trust Company savings bank brought the arrest of de la Guardia, who had protested the move. Then on May 10 the head of the National Guard, José Remón, turned against Arias, who was overthrown, and Vice President Alcibíades Arosemena assumed power on May 11. The presidential elections of 1952 made Remón president, but on Jan. 2, 1955, he was assassinated.

The year 1955 was notable for the flow of capital into Panama. Work on the Inter-American Highway and on local roads was promoted by a loan from the World Bank. Student riots against the regime of Ernesto de la Guardia (elected in 1956) and the United States led to the imposition of a state of siege in 1958. More disorders occurred during the first four months of 1959, and Independence Day generated a march, by men said to have been students, into the Canal Zone to raise the Panamanian flag. The police turned them back.

Marco A. Robles' administration

In the presidential elections of 1960, Roberto F. Chiari emerged victorious. Despite a national debt of about $83 million and a budget deficit of some $10 million, he plunged into a vast program of slum clearance, housing, hospital construction, and health service. These programs found a champion in Arnulfo Arias, whose election was thwarted by the military intimidation of the voters in 1964. The former secretary to the National Guard, Marco A. Robles, was declared the winner. Under Robles the economy of Panama was uneven. Riots in January 1964 reduced the income from the Canal Zone and frightened away foreign capital; unemployment became a serious problem. There were particularly severe disorders in March and May 1968, when Arias, a presidential candidate, with his followers in control of the National Assembly, unsuccessfully attempted the impeachment of Robles.

Arias won the election, but after 11 days in office he was removed from power by the National Guard. The guard took control of the government, and a ruling junta instituted censorship of the press, suspended constitutional guarantees, and dissolved the National Assembly.

By degrees, the junta's Colonel (later General) Omar Torrijos Herrera emerged as the leader of Panama. The constitution was again changed to strengthen and enlarge his powers. In 1972 a new national assembly of 505, selected by Torrijos, gave him full executive powers and allowed him to rule as a dictator. Behind a facade of popular government and through spectacular public works programs, the appearance of Panama City was transformed. The cost of these programs, however, plunged the country into heavy debt, and by 1977 it threatened an economic crisis. The dictatorship appointed and dismissed puppet presidents at will. In mid-1978 Torrijos' apparent triumph in obtaining U.S. approval for the Panama Canal treaties seemed to promise economic respite. (A.R.W./Gu.A.)

On July 31, 1981, Torrijos was killed in a plane crash, and command of the National Guard, by then a full-fledged army trained by the United States, went to two successive colonels. In March 1982 Colonel Rubén D. Paredes became commander of the guard; when he resigned in September 1983 to pursue the presidency, control of the guard and the country went to Colonel Manuel Antonio Noriega Moreno, former head of the secret police and a one-time operative of the U.S. Central Intelligence Agency.

Noriega and drug trafficking

Noriega consolidated the dictatorship of Torrijos. He increased the size of the army, closed every newspaper and every radio and television station that opposed his wishes, and created paramilitary regiments to brutalize and assassinate his opponents and to protect himself and his family. Customhouses, post offices, the Colón Free Zone, and other state-run enterprises were taken over by the guard. More ominous was Noriega's reported involvement in the narcotics traffic in collusion with Colombian cartels. By the spring of 1987 charges of drug-related activities, murdering of opponents, and rigging of elections were made public against Noriega by the second-in-command of the National Guard, Roberto Díaz Herrera.

In 1984 Noriega permitted the first presidential elections in 16 years. Arias was the apparent winner, but after many delays in the vote count and suspected tampering, Nicolás Ardito Barletta, Noriega's candidate, became another puppet president. Eleven months later Barletta was allegedly deposed by Noriega and replaced by the first vice president, Eric Arturo Delvalle. In February 1988, Delvalle attempted to dismiss Noriega, who was being publicly condemned by factions within and outside the country for his oppressive measures. In the United States, Noriega was indicted in the same month on counts of drug trafficking and racketeering. Delvalle's action resulted in his own dismissal, by orders of the Noriega-dominated National Assembly, and he was forced to flee Panama for his life.

U.S. President Ronald Reagan refused to recognize Delvalle's successor, and on March 11, 1988, he imposed sanctions, including the elimination of preferential trade for Panama and the withholding of canal fees. On March 16 an attempted military coup failed to overthrow Noriega, and the National Guard intensified its terroristic tactics against antigovernment demonstrators.

In the presidential elections of May 7, 1989, Guillermo Endara and his two vice presidents won by a wide margin, although they suffered brutal physical assaults by Noriega's forces, an action recorded by television cameras. Noriega then canceled the election results. Nullification of the election only exacerbated public discontent with the dictatorship, but Noriega remained impervious to criticism. On Sept. 1, 1989, he installed a classmate as president, but the failure of a bloody coup attempt against Noriega on October 3 seemed to intensify his desire to remain in power. On Dec. 15, 1989, Noriega sought and was given by the legislature the title of chief executive officer of the government. The Noriega-led assembly declared that a state of war with the United States existed. The next day Panamanian soldiers killed an unarmed U.S. Marine officer dressed in civilian clothes.

Retaliation by the United States was quick and decisive. On December 19, U.S. President George Bush ordered troops to Panama with the announced aims of seizing Noriega to face drug charges in the United States, of protecting American lives, and of restoring Panamanian liberties. Guillermo Endara and his two vice presidents were sworn in to head the government of Panama. The U.S. forces quickly quelled the resistance, but some of Noriega's soldiers carried out looting in Panama City and Colón. Noriega sought and was given refuge in the Vatican embassy in Panama, where he remained for 10 days, finally surrendering to the U.S. Army on Jan. 3, 1990. He was then transported to Miami, Fla., where he was arraigned on a host of criminal charges. In Panama Noriega was also charged with various crimes, including murder. (Gu.A.)

Treaty relations with the United States. Throughout the years of Panama's independent existence, treaty relations with the United States have been subjected to several major changes. By the protocol of 1936, the United States yielded the right of land seizure for canal purposes and the obligation to defend Panama. The worldwide depression of the 1930s and the departure from the gold standard prompted an increase in the canal annuity. In 1953 the annuity was again raised and U.S. landholdings decreased, opening the door for Panamanians to build roads across the isthmus and to manage sanitation.

In 1958 the U.S. Congress appropriated $19 million for a bridge across the canal, and an instrument of transfer was signed that conveyed to Panama real estate valued at about $25 million. The disorders in 1958–59, with the demands to fly the Panamanian flag in the Canal Zone, led to recognition by President Eisenhower that titular sovereignty over the zone resided with Panama and to the display of the flags of both nations at specified places in the zone.

On Jan. 9, 1964, U.S. and Panamanian schoolboys engaged in a scuffle over flying their flags over Balboa High School. Several thousand people turned the melee into a riot that cost the lives of about 20 persons and brought injuries to scores of others. Panama blamed the Americans, severed relations, and demanded reparations. The United States, in turn, rejected the accusations and charged Panama with inciting the riot. An International Commission of Jurists (with a minor exception) later upheld the U.S. contentions.

Treaties of 1967

Negotiations between the United States and Panama during the first part of the Robles administration led to three new protocols, signed in June 1967. The first protocol abrogated the accord of 1903, reduced the size of the Canal Zone, and provided for joint operation of the canal. The second protocol continued the responsibility of the United States for the Canal Zone's defense, and the third protocol provided for a possible sea-level canal. These proposals aroused objections from many affected quarters, and they remained unratified.

On Sept. 1, 1970, Torrijos formally notified the United States of his rejection of the agreements of 1967, but seven months later he moved to resume negotiations. Panama succeeded to one of the two hemispheric seats on the Security Council of the United Nations, and its delegate in March 1973 introduced a resolution urging continued negotiation under the auspices of the United Nations. The United States vetoed the resolution. The Panamanians continued to press their cause in negotiations that resulted in a set of principles (1974) to serve as guidelines for a new treaty; one of these was that U.S. control over canal and zone would be limited in duration.

The reaction to the proposed ultimate transfer in the U.S. Congress in 1974 and 1975 was hostile. Torrijos possessed several methods of pressure with which to respond. If there was not a peaceful settlement, he declared, then there would be violence, and this produced in Congress the abhorrent spectre of "another Vietnam." To further intimidate the opposition, Torrijos, with a considerable entourage, paid a visit to the Cuban leader Fidel Castro in January 1976.

The U.S. presidential elections of 1976 held up the treaty conversations, but after the election of Jimmy Carter progress was rapid. On Aug. 10, 1977, agreement was announced, and Carter and Torrijos signed the documents on September 7 in Washington, D.C. Despite polls that showed the treaty did not have popular support, the Senate ratified it in March 1978. The new basic treaty provided for gradual transfer of the operations of the canal to Panamanians, the phasing out of U.S. military bases, and reversion of lands and waters used in the management of the canal. Similarly, Panama was to assume jurisdiction over the zone by degrees. A second pact promised an open and neutral canal in peace and war for all nations. The transfer was to be completed by Dec. 31, 1999, but, in ratifying the treaties in March and April 1978, the Senate attached reservations that extended U.S. rights to defend the canal beyond that date and to maintain limited rights of intervention. Panama had approved the treaties in a national plebiscite in October 1977, and the ceremonies of signing were observed on June 16, 1978, in Panama City. The U.S. Congress passed legislation to implement the treaties on Sept. 26, 1979, and the treaties went into effect on Oct. 1, 1979. (A.R.W./Gu.A.)

For later developments in the history of Panama, see the *Britannica Book of the Year* section in the BRITANNICA WORLD DATA ANNUAL.

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Belize: TOM BARRY, *Belize: A Country Guide* (1989), concisely studies politics, economy, and society, as well as foreign influences on Belize. A.C.S. WRIGHT *et al., Land in British Honduras* (1959), reports on soils and agricultural land use. O. NIGEL BOLLAND and ASSAD SHOMAN, *Land in Belize, 1765–1871* (1977), studies the origins of the system of land tenure and patterns of land use and distribution. GRANT D. JONES, *The Politics of Agricultural Development in Northern British Honduras* (1971), focuses on the Corozal region between 1848 and 1968. NORMAN ASHCRAFT, *Colonialism and Underdevelopment: Processes of Political Economic Change in British Honduras* (1973), studies small farming and urban markets in central Belize in the mid-1960s. ASSAD SHOMAN, *Party Politics in Belize* (1987), is a brief examination of the political system by a participant. JULIO A. FERNANDEZ, *Belize: Case Study for Democracy in Central America* (1989), studies Belizean politics and international affairs. O. NIGEL BOLLAND, *Colonialism and Resistance in Belize* (1988), contains nine essays on aspects of social and cultural history from the early British settlement to nationhood. VIRGINIA KERNS, *Women and the Ancestors: Black Carib Kinship and Ritual* (1983), is an anthropological study.

General histories include NARDA DOBSON, *A History of Belize* (1973), from the growth of the British settlement to 1970; D.A.G. WADDELL, *British Honduras: A Historical and Contemporary Survey* (1961, reprinted 1981), including a good but dated bibliography; A.R. GREGG, *British Honduras* (1968), a popular account written for the British government; WILLIAM D. SETZEKORN, *Formerly British Honduras: A Profile of the New Nation of Belize,* rev. ed. (1981), a wide-ranging collection of information gathered on the eve of independence; and O. NIGEL BOLLAND, *Belize: A New Nation in Central America* (1986), the first general historical and contemporary survey after independence. Specific periods are addressed by O. NIGEL BOLLAND, *The Formation of a Colonial Society: Belize, from Conquest to Crown Colony* (1977), a thoroughly researched history from the Mayan period to 1871; R.A. HUMPHREYS, *The Diplomatic History of British Honduras, 1638–1901* (1961, reprinted 1981), the standard work on the colony's international relations; J. ANN ZAMMIT, *The Belize Issue* (1978), a brief summary of the dispute with Guatemala; WAYNE M. CLEGERN, *British Honduras: Colonial Dead End, 1859–1900* (1967), a well-documented study of economic and political changes in the late 19th century; and C.H. GRANT, *The Making of Modern Belize: Politics, Society, & British Colonialism in Central America* (1976), which studies Belize between 1950 and 1974. RALPH LEE WOODWARD, JR., *Belize* (1980), is an annotated bibliography on all aspects of the country.

Costa Rica: HAROLD D. NELSON (ed.), *Costa Rica: A Country Study* (1984); and TOM BARRY, *Costa Rica: A Country Guide* (1989), provide general information on Costa Rican geography, society, economy, and history. CAROLYN HALL, *Costa Rica: A Geographical Interpretation in Historical Perspective* (1985), offers comprehensive and up-to-date information on Costa Rican geography. JOHN BIESANZ and MAVIS BIESANZ, *Costa Rican Life* (1944, reprinted 1979); and RICHARD BIESANZ, KAREN ZUBRIS BIESANZ, and MAVIS BIESANZ, *The Costa Ricans* (1982), examine Costa Rican society, history, and institutions. MITCHELL A. SELIGSON, *Peasants of Costa Rica and the Development of Agrarian Capitalism* (1980), reveals much about rural life in the country. LOWELL GUDMUNDSON, *Costa Rica Before Coffee: Society and Economy on the Eve of the Export Boom* (1986), is an excellent reevaluation of the early development of Costa Rica and includes considerable discussion of the myths of Costa Rican history and its historiography. ANTHONY WINSON, *Coffee and Democracy in Modern Costa Rica* (1989), analyzes the relationship of small-farmer coffee production and the development of democracy. CHARLES D. AMERINGER, *Democracy in Costa Rica* (1982), is the best source available in English on the formation and persistence of democracy in Costa Rica. MARILYN APRIL DORN, *The Administrative Partitioning of Costa Rica: Politics and Planners* (1989), emphasizes the lack of consistency among political (province) and administrative regionalization schemes.

CHESTER LLOYD JONES, *Costa Rica and Civilization in the Caribbean* (1935, reprinted 1967), is dated but excellent for the early national history of the country. ADELA F. DE SÁENZ and CARLOS MELÉNDEZ CH., *Nueva historia de Costa Rica,* 5th ed. (1982), provides an excellent broad survey. MARIO RODRÍGUEZ, *A Palmerstonian Diplomat in Central America* (1964), thoroughly examines the life of Frederick Chatfield, who made Costa Rica into a virtual British protectorate. WATT STEWART, *Keith and Costa Rica* (1964), discusses the fascinating career of the railroad builder, financier, and fruit company executive. JOHN PATRICK BELL, *Crisis in Costa Rica* (1971), is the definitive investigation of the 1948 revolution. CHARLES D. AMERINGER, *Don Pepe* (1978), is a political study of a major Costa Rican figure in the 20th century, José Figueres. JOSÉ LUIS VEGA CARBALLO, *Orden y progreso: La formación del estado nacional en Costa Rica* (1981), objectively surveys social and political change in Costa Rica.

El Salvador: Physical geography is graphically presented in INSTITUTO GEOGRÁFICO NACIONAL INGENIERO PABLO ARNOLDO GUZMÁN, *Atlas de El Salvador,* 3rd ed. (1979). ALISTAIR WHITE, *El Salvador* (1973, reissued 1982), provides a comprehensive analysis of social and economic development. DAVID BROWNING, *El Salvador: Landscape and Society* (1971), deals with the population as well as the country, focusing on "man-land" relationships.

PHILIP L. RUSSELL, *El Salvador in Crisis* (1984), is the best choice for an overview. Of the many general studies that centre on the period since 1931, some of the most useful include ENRIQUE A. BALOYRA, *El Salvador in Transition* (1982), a penetrating analysis of the political system; RAYMOND BONNER, *Weakness and Deceit: U.S. Policy and El Salvador* (1984), a critical account by a former *New York Times* correspondent; JAMES DUNKERLEY, *The Long War: Dictatorship and Revolution in El Salvador,* new ed. (1985), a scholarly, critical examination of events leading to the civil war; TOMMIE SUE MONTGOMERY, *Revolution in El Salvador: Origins and Evolution* (1982), an informed commentary on the role of the Roman Catholic church and the rise of the left; and MICHAEL MCCLINTOCK, *The American Connection* (1985), a detailed discussion of the U.S. role in the Salvadoran counterinsurgency effort. THOMAS P.

ANDERSON, *Matanza: El Salvador's Communist Revolt of 1932* (1971), stands as the definitive study of the peasant uprising. Anderson's *The War of the Dispossessed: Honduras and El Salvador, 1969* (1981), is an account of the Soccer War, including his firsthand observations; while WILLIAM H. DURHAM, *Scarcity and Survival in Central America* (1979), provides insightful treatment of the ecological and socioeconomic factors that led to that war. PATRICIA PARKMAN, *Nonviolent Insurrection in El Salvador* (1988), explains the fall of the Hernández Martínez dictatorship in 1944. STEPHEN WEBRE, *José Napoleón Duarte and the Christian Democratic Party in Salvadoran Politics, 1960–1972* (1979), focuses on a critical period in El Salvadoran history. RALPH LEE WOODWARD, JR., *El Salvador* (1988), offers the best bibliographic guide to further study.

Guatemala: FRANCIS GALL (comp.), *Diccionario geográfico de Guatemala,* 2nd ed., 4 vol. (1976–83), is a comprehensive work. General geographic aspects are covered in RICHARD F. NYROP (ed.), *Guatemala: A Country Study,* 2nd ed. (1983). Economic and political studies include J.C. CAMBRANES, *Coffee and Peasants: The Origins of the Modern Plantation Economy in Guatemala, 1853–1897* (1985); and PETER CALVERT, *Guatemala: A Nation in Turmoil* (1985). SEVÉRO MARTÍNEZ PELÁEZ, *La patria del criollo* (1970, reissued 1985), describes in detail the colonial social structure and the formation of the Creole mentality.

A good general history is JIM HANDY, *Gift of the Devil: A History of Guatemala* (1984). CHESTER LLOYD JONES, *Guatemala, Past and Present* (1940, reissued 1966), is still useful. THOMAS MELVILLE and MARJORIE MELVILLE, *Guatemala: The Politics of Land Ownership* (1971), discusses the question of landholding and the lack of reform. The violence that has characterized the history of Guatemala since 1945 is detailed in EDUARDO GALEANO, *Guatemala: Occupied Country* (1969); SUSANNE JONAS and DAVID TOBIS (eds.), *Guatemala* (1974); ROGER PLANT, *Guatemala: Unnatural Disaster* (1978); and JONATHAN L. FRIED, *et al.* (eds.), *Guatemala in Rebellion: Unfinished History* (1983), an Americas Watch Committee report.

Honduras: JAMES D. RUDOLPH (ed.), *Honduras: A Country Study,* 2nd ed. (1984), covers the geography, social and administrative structure, and history of the country. Political aspects of resources are discussed in KENNETH V. FINNEY, "Rosario and the Election of 1887: The Political Economy of Mining in Honduras," *The Hispanic American Historical Review,* 59(1):81–107 (February 1979); and JAMES A. MORRIS and STEVE C. ROPP, "Corporatism and Dependent Development: A Honduran Case Study," *Latin American Research Review,* 12(2):27–68 (Summer 1977).

Works on various aspects of Honduran history include ROBERT S. CHAMBERLAIN, *The Conquest and Colonization of Honduras, 1502–1550* (1953, reprinted 1966); NANCY PECKENHAM and ANNIE STREET (eds.), *Honduras: Portrait of a Captive Nation* (1985); and JAMES A. MORRIS, *Honduras: Caudillo Politics and Military Rulers* (1984). WILLIAM H. DURHAM, *Scarcity and Survival in Central America* (1979), pursues the underlying causes of the 1969 Soccer War between El Salvador and Honduras and exposes many of the socioeconomic problems of Central America and their long-term historical consequences. THOMAS P. ANDERSON, *The War of the Dispossessed* (1981), is another account of that war.

Nicaragua: JAMES D. RUDOLPH (ed.), *Nicaragua: A Country Study* (1982), is a good general introduction. On the country's geography, see JAIME INCER, *Geografía ilustrada de Nicaragua* (1973). Political and social studies include BERNARD DIEDERICH, *Somoza and the Legacy of U.S Involvement in Central America* (1981); BERNARD NIETSCHMANN, *The Unknown War: The Miskito Nation, Nicaragua, and the United States* (1989); and DAVID CLOSE, *Nicaragua: Politics, Economics, and Society* (1988). The relationship between politics and the church is addressed in MICHAEL DODSON and LAURA NUZZI O'SHAUGHNESSY, *Nicaragua's Other Revolution: Religious Faith and Political Struggle* (1990).

General works on Nicaraguan history and politics include JOHN A. BOOTH, *The End and the Beginning: The Nicaraguan Revolution,* 2nd ed., rev. and updated (1985); and THOMAS W. WALKER, *Nicaragua: The Land of Sandino,* 2nd ed., rev. and updated (1986). KARL BERMANN, *Under the Big Stick: Nicaragua and the United States Since 1848* (1986), provides a solid overview of U.S. involvement in Nicaraguan affairs since the mid-19th century. Two classic studies of important aspects of the country's prerevolutionary 20th-century history are NEILL MACAULEY, *The Sandino Affair* (1967, reprinted 1985); and RICHARD MILLETT, *Guardians of the Dynasty* (1977), which traces the history of the U.S.-created National Guard and its relationship with the Somoza family. DONALD C. HODGES, *Intellectual Foundations of the Nicaraguan Revolution* (1986), provides useful insights into both Augusto César Sandino and the Sandinistas. An informed critique of Nicaragua's revolutionary elite is DENNIS GILBERT, *Sandinistas: The Party and the Revolution* (1988). Among numerous useful studies of the conflict between the administration of U.S. President Ronald W. Reagan and Sandinista Nicaragua are PETER KORNBLUH, *Nicaragua: The Price of Intervention: Reagan's Wars Against the Sandinistas* (1987); and THOMAS W. WALKER (ed.), *Reagan Versus the Sandinistas: The Undeclared War on Nicaragua* (1987). For insight into specific aspects of the Nicaraguan revolution, see ROBERT F. ARNOVE, *Education and Revolution in Nicaragua* (1986); MARY B. VANDERLAAN, *Revolution and Foreign Policy in Nicaragua* (1986); and GARY RUCHWARGER, *People in Power: Forging a Grassroots Democracy in Nicaragua* (1987). RALPH LEE WOODWARD, JR. (comp.), *Nicaragua* (1983), is an annotated bibliography.

Panama: SANDRA W. MEDITZ and DENNIS M. HANRATTY (eds.), *Panama: A Country Study,* 4th ed. (1989), provides information on all aspects of the country. Physical geography is covered by ROBERT A. TERRY, *A Geological Reconnaissance of Panama* (1956). INSTITUTO GEOGRÁFICO NACIONAL TOMMY GUARDIA, *Atlas nacional de la República de Panamá,* 3rd ed. (1988), graphically presents information on the geography, resources, economy, social conditions, and administrative and political divisions. BURTON L. GORDON, *A Panama Forest and Shore: Natural History and the Amerindian Culture in Bocas del Toro* (1982), is also useful. The role of agriculture in the country's economy is presented by WILLIAM C. MERRILL *et al., Panama's Economic Development* (1975). STEVE C. ROPP, *Panamanian Politics: From Guarded Nation to National Guard* (1982), discusses Panama's politics and foreign relations.

Works on Panama's history include JOHN EASTER MINTER, *The Chagres: River of Westward Passage* (1948); LAWRENCE O. EALY, *Yanqui Politics and the Isthmian Canal* (1971); GUSTAVE ANGUIZOLA, *Philippe Bunau-Varilla: The Man Behind the Panama Canal* (1980); STEPHEN G. RABE, *Eisenhower and Latin America: The Foreign Policy of Anti-communism* (1988); and LESTER D. LANGLEY, *The United States and the Caribbean in the Twentieth Century,* 4th ed. (1989).

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O.H.H./T.P.A./ B.Ni./T.W.Wa./B.L.G./Gu.A.)

Central Asian Arts

The term Central Asia has no specific boundaries in a geographical context, but cultural historians generally apply it to the territory bordered on the north by Siberia in the Soviet Union and on the south by Iran, Pakistan, Bangladesh, and India. This area includes the present Soviet republics of Kazakhstan, Turkmenistan, Uzbekistan, Kirgizstan, and Tadzhikistan and portions of the Sinkiang Uighur Autonomous Region of China—the populations of which are largely of Turkic stock—and the kingdom of Afghanistan. For the purposes of this article, the term is further extended to include Siberia and the Soviet Arctic lands and, in the south and east, the Altaic regions of Mongolia and Manchuria (Northeast Provinces) and the Himalayan regions of Nepal, Bhutan, Sikkim, and Tibet.

This immense tract of land—with its highly varied topography and climate and its diversity of ethnic and linguistic backgrounds—encouraged the development of greatly varied artistic styles and traditions among the inhabitants of widely separated regions. These differences were magnified by the emergence of dissimilar religions, which in turn encouraged the formation of distinctive schools or traditions of art. Further artistic variances can be attributed to cultural time lag, for comparable stages of artistic development were not reached simultaneously throughout the area.

The arts that developed across Central Asia often fed or were fed by those of adjoining cultural regions or by such supraregional influences as Islām. Although reference will be made to such cross-cultural interactions wherever appropriate, more detailed information on these other areas may be found in such articles as ISLĀMIC ARTS and SOUTH ASIAN ARTS. (The peoples and traditional cultures of the region are treated in the *Macropædia* articles ASIA and UNION OF SOVIET SOCIALIST REPUBLICS.) (T.T.R.)

This article is divided into the following major sections:

Literature

Of the relatively few Central Asian languages that have developed written literatures, the most important are Turkic, Tibetan, and Mongol. (For a treatment of the Central Asian languages, see LANGUAGES OF THE WORLD.) This article will deal with Tibetan and Mongolian literatures from their inception to the 20th century and with Turkic literature from its inception to the 11th century, when the Muslim invasion introduced a period of Islāmic culture. Subsequent Turkic literature is treated in the Turkish literature section of ISLĀMIC ARTS, which also treats Central Asian literatures written in Arabic and Persian. (For literature in Chinese, see the article CHINESE LITERATURE.)

TURKISH LITERATURE

The purely Turkish period in the history of Turkish literatures came before the conversion of the Turks to Islām and covers approximately the 8th to the 11th century AD. The oldest literary legacy of the period is found in the Orhon inscriptions, found in the Orhon Valley, northern Mongolia, in 1889 and deciphered in 1893 by the Danish philologist Vilhelm Thomsen. The inscriptions are on two large monuments, erected in 732 and 735 in honour of the Turkish prince Kül and his brother Bilge Kagan; they are carved in a script used also for inscriptions found in Mongolia, Siberia, and western Turkistan and called by Thomsen "Turkish runes." They relate in epic and forceful language the origins of the Turks, their golden age, their subjugation by the Chinese, and their liberation by Bilge Kagan. The polished style suggests considerable earlier development of the language. Excavations in Chinese Turkistan have brought to light specimens of writings of the Uighur Turks from the 9th to the 11th century. Maḥmūd Kāshgarī's comprehensive dictionary (1071?) contains specimens of old Turkish poetry in the typical form of quatrains (*dörtlük*), representing all the principal genres: epic, pastoral, didactic, lyric, and elegiac. (F.I.)

"Turkish runes"

TIBETAN LITERATURE

Tibetan was developed as a literary language from the 7th century onward as a result of earlier cultural contacts with neighbouring Buddhist countries, namely, the small states of the Takla Makan, especially Khotan (Ho-t'ien) and the kingdoms of ancient northwestern India (modern Gilgit, Kashmir, and Kulu) and Nepal. Scripts of Indian origin were in use in these countries, so the Tibetans also adapted an Indian script to suit their own very different language. By far the greater number of works produced between the 7th and 13th centuries are skillful translations of Buddhist works, largely from Sanskrit, on which Indian scholars and Tibetan translators worked side by side. The Tibetans had to create an entirely new (and therefore artificial) vocabulary of religious and philosophical terms, mainly by ingenious compounding of simple terms available in their own language. Apart from some religious terms in daily use, this vocabulary remains a specialized scholarly language. An indigenous literature was also produced: annals and chronicles, sets of spells and prognostications, legendary and liturgical works, all representing the remains of ancient oral traditions. Large collections of such manuscript fragments, all earlier than the 11th century, were discovered early in the 20th century in the Cave of the Thousand Buddhas near Tun-huang (at the eastern side of the Takla Makan).

Indigenous literature of Tibet

The quasi-official work of translating authorized Indian Buddhist texts, which continued for six centuries, gave incentive to the Bon-pos (the followers of the pre-Buddhist religion of Tibet) to collect and write down their own early traditions; but in so doing they adopted many Buddhist ideas and, inevitably, used the new vocabulary. The followers of the earliest Buddhist traditions to enter Tibet (the rNying-ma-pa, or "Old Order") also committed their teachings to writing; and, conversely, these are interspersed with pre-Buddhist traditions.

The official Tibetan Buddhist canon was closed in the 13th century; it consisted of two parts, the *Kanjur* ("translated word," teachings or reputed teachings of the Buddhas themselves) and the *Tanjur* ("translated treatises," mainly commentaries by Indian teachers). By this time, however, there already existed some orthodox Buddhist works of Tibetan origin (for example, *Mi-la ras-pa* and *sGam-po-pa*); and from the 13th century onward, under the impetus given by the prolixity of religious houses and orders, there were produced such lengthy and numerous collections of historical and biographical works, treatises and commentaries, liturgy and religious drama that Tibetan literature must be one of the most extensive in the world. Just as in the European Middle Ages there was little secular literature worth the name, so there is none in Tibetan except for a great epic (*Rgal-po Ge-sar dgra'dul gyi rtogs-pa brjod-pa,* "The Great Deeds of King Gesar, Destroyer of Enemies") that recounts the exploits of the king and magic hero Gesar. This work grew through the centuries, assimilating whatever material pleased the fancy of the bards.

Secular literature

After the craft of printing from incised wood blocks was introduced from China, possibly in the 14th century, certain monasteries became famous printing houses. This form of printing continued until the Chinese invasion in 1959. Manuscripts and block-printed books are always of elongated shape, thus imitating the form of ancient Indian palm-leaf manuscripts. There are considerable collections in some European libraries—London, Paris, and Rome—but few translations are available because of the small number of scholars of Tibet.

Despite the phonetic changes in the spoken dialects since the script was fixed, the Tibetans have never changed their system of writing. Thus, once the literary language and the various types of script have been mastered, the reader has immediate access to all literature of the 7th to the 20th century, though changes in style and vocabulary have left many obscurities in the earliest works. Since there is no modern style of writing, the 20th-century colloquial language can be written only in the traditional medium (as though, for example, one had to write modern Italian with Latin spellings and grammatical forms); the Tibetans themselves compose even personal letters in a conventional literary style. (D.L.S.)

MONGOLIAN LITERATURE

Mongolian literature begins with the *Secret History of the Mongols,* an Imperial chronicle dealing with the life and times of Genghis Khan and his successors, written about 1240. *Üligers,* orally transmitted epic stories in verse, form the bulk of native literary expression. Highly stylized, these tales relate adventures of legendary heroes and villains. In spite of their great length (sometimes more than 20,000 lines), they are recited from memory by bards. Like the other epics, such as the Greek *Iliad* and the Roman *Aeneid,* Mongolian epics are genuine artistic creations. The verses alliterate in couplets or quatrains, are seven or eight syllables long, and are characterized by parallelism. In addition to Genghis Khan, the epic stories tell of heroes such as Erintsen Mergen, Engke Bolod Khan, and Gesar Khan (the last of Tibetan origin). The villain of the epics is the many-headed monster, the *manggus,* whom the hero always defeats in the end.

Mongolian historical chronicles

Historical chronicles represent another important form of indigenous literature. Usually beginning with the creation of the world from primordial elements, they attempt to link the Indian and Tibetan rulers with the house of Genghis Khan. Such are the *Altan tobchi* ("The Golden Button"), composed about 1655 and giving a world history down to Ligdan Khan (1604–34); another *Altan tobchi* (written about 100 years later); and the *Erdeni-yin tobchi* ("The Jeweled Button"), written in 1662 by Saghang Sechen.

Mongolian written literature was profoundly influenced by the introduction of Buddhism around the end of the 1500s. Earlier surviving written works (on stone or paper) are mostly official documents; and no oral epics were written down until the late 1880s. The advent of Buddhism evoked translations of its sacred writings and related works. A Buddhist canonical collection, the *Bka'-'gyur* (or *Kanjur;* comprising the *sūtra*s and *vinaya* of the *Tripiṭaka*), was translated and printed in 1635 in 108 volumes; the *Bstan-'gyur* (*Tanjur*), containing canonical commentary and noncanonical works in its 225 volumes, followed in 1741. Two especially well-known sermons (*sūtras*) of Buddha are the *Altan gerel* ("Golden Beam") and the *Chagan lingqua* ("White Lotus"), or, as they are known in Sanskrit, *Suvarṇaprabhāsa* and *Saddharmapuṇḍarīka.* None of these works, however, is indigenous.

Religious but nondogmatic birth stories (called in Sanskrit *Jātaka*) deal with Buddha's meritorious deeds, and, like the parables of Christianity, illustrate religious truths. Best known is the *Üliger-Ün dalai* ("The Sea of Stories"). Translations of other Indian fables are the *Siddhi Kür* ("Tales of the Vampire") and the *Bigarmijid* ("Saga of King Vikramāditya").

In the 18th and 19th centuries Chinese traders brought from China many Mongolian translations of Chinese novels of enchantment and romance, including the *San Kuo chih yen-i (Romance of the Three Kingdoms)* and others.

20th-century literature

In the early 20th century T. Zhamtsarano, a Russian-educated Buryat writer and intellectual, founded the short-lived Mongolian newspaper *Shine toli* ("The New Mirror"). He also translated the works of some Western authors, such as Edgar Allan Poe, Jules Verne, and H.G. Wells.

After a Soviet state was established in Mongolia in the early 1920s, the power and influence of the Tibetan Buddhist church declined, and with it the literature it sponsored. There began to arise a revolutionary and Socialist literature serving the people according to the current beliefs of the Communist Party. Popular themes were social criticism of the feudal past, with its exploitation of the people, and exemplary tales that showed new Socialist values and educated readers to overcome the resistance of reactionary forces. Mongolian reliance on and gratitude for the fraternal Soviet Union and its assistance is an ever-present undercurrent. The rise of a system of schooling, extending to the university level, has enabled many young Mongolian writers, poets, artists, and actors to achieve, within limits, a self-expression never before available to them. Leading 20th-century modern writers are Dashdorjiyn Natsagorj, Ts. Damdinsürün, D. Sengee, S. Erdene, and others. (J.R.Kr.)

Music

Music in Central Asia flowered along centuries-old caravan routes linking the Near East with China and India via Turkistan, the vast region extending from the Caspian Sea to Sinkiang province in China. Musical instruments diffused from one region to another, and many of the musical styles still display foreign influence. The variety of musical styles ranges from the systematically organized classical music of Turkistan, to the notated religious chants of Tibet, to the highly varied folk music styles of the region's numerous ethnic groups. The main thrust of this examination of Central Asian music will be on the traditions and styles first of Afghanistan and the sedentary population of Turkistan, then of the Turkic nomads, the Mongols, and the Siberian peoples, and, finally, of the Himalayan peoples in Tibet, Bhutan, Nepal, and Sikkim.

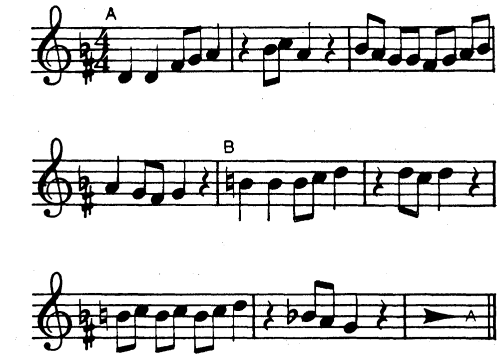
AFGHANISTAN AND THE SEDENTARY POPULATION OF TURKISTAN

This region of Central Asia includes Afghanistan, present-day Uzbek Soviet Socialist Republic (Uzbekistan) and Tadzhik S.S.R. (Tadzhikistan), and the oases of eastern (Chinese) Turkistan. The region lies within the Persian cultural area, and in the arts and in language the Persian imprint has endured over many centuries. In music the links with Persia appear most clearly in terminology and instruments. Islām, another Near Eastern heritage, predominates in this region and results in a generally low social status of musicians and musical performance—a situation generally not found in other regions of Central Asia.

The area includes two main streams of musical practice: folk music in a broad range of styles, often closely linked to specific ethnic groups; and the more exclusive, cosmopolitan, classical music, derived from the medieval court music of Bukhara, Samarkand, and other urban centres of Transoxania (modern Uzbek S.S.R., Tadzhik S.S.R., and Turkmen S.S.R.). A third stream is now in the process of formation: popular music disseminated through the mass media.

Musical instruments

Folk music. Generally characterized by a scarcity of musicians and musical instruments, folk music of this region is predominantly a matter of solo playing and singing, small ensembles, and a complete lack of musical notation or codified musical theory. In their general types, the musical instruments are closely related to those of Persia and the Near East, but specific forms and playing styles are purely local. Thus, there are numerous variants of the Persian long-necked lute, with names derived from the Persian *tanbūr* or *dūtār;* small spike fiddles, in which the neck skewers the body, forming a spike at the base; various block or fipple flutes, with air ducts like that of the Western recorder; transverse (horizontally held) flutes; oboes; metal jew's harps; and two basic drum types, a single-headed vase-shaped drum of pottery or wood and a large single-headed frame drum, or tambourine—all instrumental types widely diffused in the Near East.

Stylistically, the music relates to that of both the Near East and the surrounding nomadic Turkic peoples of Central Asia. Songs are completely monophonic (*i.e.,* consisting only of a line of melody), but instrumental music often includes two-part polyphony (music in more than one voice, or part). The polyphony may take the form of a drone (sustained note) with a melody played above it. Or it may be organum style; *i.e.,* the second part playing the same melody as the first but at a higher or lower pitch. Most common are parallel fourths or fifths (a fourth encompasses four notes of a Western major or minor scale; a fifth, five). In structure, much of the music is based on small forms, frequently binary, or two-section, and ternary, or three-section. Small musical units may be repeated many times and varied slightly at each appearance. The recurrence of melodic phrases and an emphasis on marked rhythms is common and is related to the frequent role of music as dance accompaniment. Thus in the following example, a dance tune from Afghan Turkistan, sections A and B are similar in their overall melodic structure, and also in the small units of three or four notes on which they are built. The sequence ABAB, etc., is repeated throughout the dance. Vocal music may have greater rhythmic flexibility and melodic range, but in form it is almost always subordinated to the structure of the song text. Quatrains such as the *rubai* and *charbaitai* are the most prominent village verse forms, with the exception of the *lundai,* a couplet used by the nomadic Pashtuns of Afghanistan. In the urban oases, couplet forms based on the classical Persian *ghazal,* a lyric poem of 6 to 15 couplets, are more common.

Attitudes toward music

The generally negative attitude of Islām toward music has led in Afghanistan to strictures against musical performance and to extremely low social status for musicians. Music is heard mainly in male-dominated public teahouses or at private celebrations such as weddings and circumcisions. Women may have their own musical genres within their enclosures; in this context the strong tradition of women's music in the Fergana Valley of the Uzbek S.S.R. is noteworthy.

Within this general picture there is enormous diversity. The Uzbeks (a Turkic people) and Tadzhiks (an Iranian group), who live side by side across northern Afghanistan and southern Turkistan, tend to share many musical traits and instruments. In contrast, most groups, such as the Pashtuns, Ḥazāra, and Baluchs of Afghanistan and Pakistan or, in the extreme, the isolated mountain peoples of Nūrestān in Afghanistan and of the Pamirs in the Tadzhik S.S.R. and Afghanistan, have maintained distinctive musical styles and, in some cases, unique musical instruments. The Nūrestāni *vaji,* an arched harp (having a bow-shaped body with no forepillar), is a striking case of the possible survival of an instrument type on the margins of a now disintegrated culture area; there are no other harp traditions today between the Caucasus and Burma, although iconographic evidence indicates that in ancient times harps were widespread in Central Asia, the Near East, and India. The music of the Ḥazāra includes vocal effects produced by striking the throat while singing, causing a break in the sound, and Baluchi music also features a broken-voice style.

Medieval court music

Classical music. In contrast to the folk music styles just described, the court-derived classical style of Bukhara and Samarkand represents a highly systematic, theoretically grounded, cosmopolitan musical tradition. Lying along the medieval silk route, the Turkistani oases were open to musical cross-currents. Today's musical roots may reach back to the period in which urban Central Asian music was in vogue at T'ang dynasty courts in China (618–907). The movement of musical instruments across the caravan trail from the Near East to China via Central Asia is well documented since early times. Over the centuries, town musicians evolved an urban style patronized by the local courts, notably under Timur (Tamerlane) and his descendants (*c.* 1350–1500) in Herāt (now in Afghanistan) and Samarkand. The degree of musical eclecticism characteristic of the era is illustrated by a court historian's description of the festivities of Timur's son:

> Golden-tongued singers and sweet-sounding musicians played and sang to motives [melodic figures] in Persian style, to Arab melodies according to Turkish practice and with Mongol voices, following Chinese laws of singing and Altai meters.

By the 17th century the court style had been codified into sets of nonimprovised suites of instrumental and vocal pieces using poetic texts in classical Persian and local court Turkish (Chaghatai). In Bukhara this collection of suites was known as the *Shashmaqām,* or six *maqām*s (suites), with each *maqām* (an Arabic term, but changed in meaning) set in one of the classical Persian musical modes. (The Persian modes are melodic frameworks, each with a given scale, typical melodic figures, and accepted emotional content.) Regional courts and large towns developed their own sets of *maqām*s, performed in unison by an orchestra and a male chorus.

20th-century changes in music practice

Areas of Turkistan under Soviet rule since about 1920 have undergone far-reaching modification of traditional music practice, although the older styles such as the *Shashmaqām* are also maintained. Changes include the reconstruction of local instruments to fit the Western musical scale of 12 equally spaced half steps, establishment of music schools and conservatories, creation of orchestras of folk instruments, introduction of vocal polyphony, and the writing of works in Western forms (symphonies, operas, chamber music) by native and European Soviet composers. In Afghanistan, musical change began on a national basis in the 1950s under the influence of Radio Afghanistan, which broadcasts principally popular styles based on Pashtun folk music and songs of the Indian film industry.

TURKIC NOMADS, MONGOLS, AND SIBERIAN PEOPLES

This region includes primarily the great open spaces of Central Asia, from the Turkmen desert in the southwest to the Kazakh steppes, Mongol plains, and from the Gobi to the vast subarctic Siberian evergreen forests, or taiga, and tundra, or Arctic plains, stretching to the Pacific. The considerable mobility and often close linguistic affinity of the peoples in the area led to substantial interchange of musical terms and instruments, and to common social functions of music relating to the traditional tribal social structure of most of the groups of this region.

Social role of music. Three basic functions of music are common throughout most of the region: music as ritual, with magical connotations (shamanism), music as tribal record, aiding group solidarity (epic recitation), and music as entertainment (itinerant performers, festivals).

The shaman's music

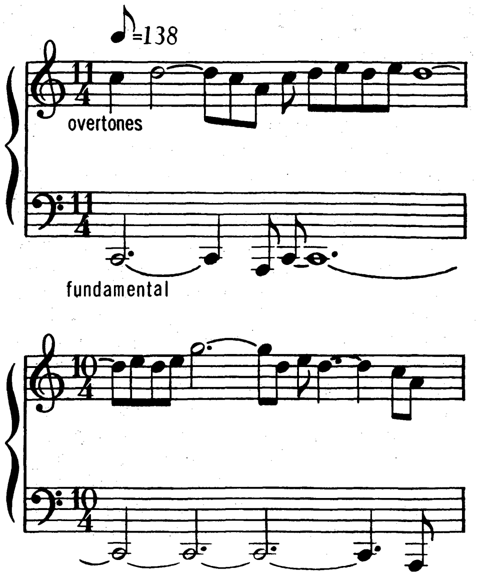
Music is the medium of the shaman, or priest-medicine man, as he performs his role as mediator between the seen, or men, and the unseen, the spirits that inhabit the spheres above and below the earth. Traditional shamanistic séances were creative, impassioned musico-dramatic scenes produced by a single performer, the shaman. Not only is music the shaman's aid in inducing the trance that enables him to contact spirits, but in Siberia his drum (a very large tambourine) may be considered a steed for the trip to other worlds. Thus great attention is given to each stage of drum construction, from selecting the wood of certain trees to the painting of symbolically charged designs on the drumhead. The metal hangings, sometimes including bells, on the shaman's costume also play a musical role. Among the Kirgiz and Kazakhs and until recently among the Turkmens, a fiddle with horsehair strings and bow performs the same function as the Siberian drum. Metal ringlets are attached to the head of the fiddle, and a niche is hollowed there for a mirror to catch the reflections of spirits. Shamans' horsehair fiddles can even be found among townspeople of northern Afghanistan. The occurrence of shamanism has sharply declined in Soviet portions of Central Asia.

Epic recitation, which may serve as tribal history, also has magical overtones. Among Turks, the same term (*bakhshi*) may be used for both shamans and bards, and both may be called to their trade by spirits to undergo a difficult period of initiation. Storytellers use a fiddle or lute as accompaniment, and tales may run through several nights of exhaustive performance; one Kirgiz bard is known to recite 300,000 verses of the *Manas,* the major Kirgiz epic. Such marathon performances are facilitated by the use of stereotyped melodic motives—standard short melodic figures—often invented by the individual performer. Local epic traditions vary widely in dramatization; *i.e.,* the proportion of dialogue, monologue, and narrative.

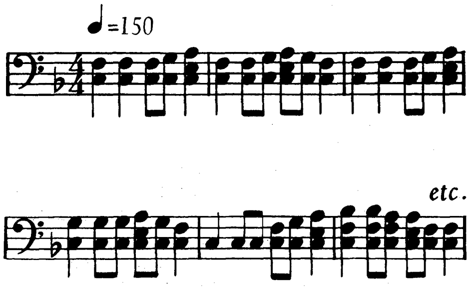
The singing contest

The third area-wide musical function, entertainment, takes many forms. One common diversion is the singing contest, in which rival minstrels compete in wit and virtuosity. Such trials of skill are most notable among the Kirgiz, Kazakhs, and Mongols. The contests follow strict rules of versification, musicality, and procedure. Often the loser must pay a forfeit to the victor, who receives acclaim from the audience and gifts from wealthy patrons; a singer's reputation may be made or broken in a single afternoon. Frequently a contestant will vilify the clan championed by the opposing singer and laud his own faction. In Siberia another type of entertainment is the widespread practice of bear festivals at specific times of the year, during which a bear is killed and his head displayed, to the accompaniment of music, dance, and games.

Instrumental and vocal styles. Across the region the principal instrument types are lutes, with two or three strings, the necks either fretted or fretless; fiddles, largely horsehair fiddles; flutes, mostly open at both ends and either end-blown or side-blown; and jews' harps, either metal or, often in Siberia, wooden. Few percussion instruments are found, except for the shaman's magic drum. Considerable instrumental polyphony is played on lutes and fiddles, particularly among the Turkic peoples. Vocal polyphony may occur in special ways: singers among the Mongols and Tuvins (a Siberian people northwest of Mongolia) can produce two parts while singing solo, as in the example below, by strongly reinforcing upper partials (overtones) while singing a very deep fundamental pitch. West of the Urals, Bashkirs may hum a basic pitch while playing solo flute pieces, and certain Siberian peoples may sing choral overlapping responsorial songs (in which group and soloist alternate, one beginning slightly before the other finishes).



The vast geographic stretch of the region produces musical links to neighbouring areas as well as highly distinctive local styles. The Turkmens live in Afghanistan and Iran as well as Turkistan and manifest some Persian influence in musical terms and instruments, yet they possess unique vocal and instrumental styles. Particularly striking is their series of guttural sounds serving as vocal ornaments. The Kirgiz and Kazakhs, closely related musically, maintain ties to Mongol and northern styles (*e.g.,* of the Bashkir and Tatar peoples, west of the Urals) as well as to Turkistan. Nevertheless, their relaxed voice quality, musical scales, and distinctive instrumental polyphony set them off. Noteworthy here is the versatile polyphonic style of the three-stringed Kirgiz *komuz* lute, based on extensive development of short melodies called kernel tunes. In the *komuz* piece shown below, the kernel tune is stated in the first two measures and is varied and developed elaborately as the piece progresses. Another Kirgiz-Kazakh specialty is programmatic music, in which instrumentalists suggest situations or tell specific stories without words, through musical images alone.



Mongolian and Siberian music

The Mongols display links to both Chinese and Tibetan music. Chinese influence is apparent in the use of certain instruments (*e.g.,* some flutes and fiddles) and perhaps in the structure of melodies; Tibetan impact occurs in the religious music and musical instruments of Tibetan Buddhism, introduced in the 16th century. Mongolian music also has its own distinctive profile, sporadically documented since the 13th-century *Secret History of the Mongols,* the first written Mongolian chronicle. Of interest is the fact that Arghūn Khān, Mongol ruler of Persia, sent a musician as emissary to Philip IV the Fair of France in 1289. Because of the focal position of Mongolia at the heart of Central Asia, some Mongol epic melodies have spread westward as far as the Kalmuks on the Volga River, and eastward to the Ainu of Sakhalin Island, north of Japan. Mongol songs may be either quick and marked rhythmically or drawn-out in free rhythm, with extensive melodic ornamentation. The Mongol horsehair fiddle accompanies a singer with simultaneous variations on the melody, a technique called heterophony.

Siberian music includes a broad spectrum of styles over a huge geographic expanse. Many unique traditions occur, such as the bridgeless, often rectangular zithers of the Khants and Mansi, Ugrian peoples along the Ob River; farther east, the solo flute-and-voice polyphony of the Tuvins and Bashkirs; and the rapid, compact songs with nonsense syllables of the Gilyaks, Chukots, and other peoples of the Far Eastern Amur River region and Pacific coast. At that northeastern shore of Siberia there is a carryover of musical style to the Ainu of northeastern Japan and possible musical ties between the Eskimos of Asia and of North America. Other links beyond Central Asia may exist at the far western end of Siberia, for example, to the music of Lapland in the Scandinavian Arctic; or the relation of tunes of certain peoples of the Volga River region, such as the Mari or Cheremis, with old Hungarian folk songs.

Outside the few written Mongol references to music, the only approach to discovery of the stylistic history of this region of Central Asia is through fragmentary information about musical instrument types. Perhaps the most remarkable instrument finds were made at Pazyryk in south central Siberia, where Soviet archaeologists found wooden objects which possibly form pieces of a harp and an artifact resembling a vase-shaped drum, both dating from the 5th century BC.

THE HIMALAYAN PEOPLES

This region, including Tibet, Bhutan, Nepal, and Sikkim (which was annexed to India in 1975), occupies an important middle ground between India and China, and its central position is reflected in the local music cultures. Of utmost importance for musical life was the introduction of Buddhism from India via Turkistan, beginning in the 7th century AD. Music became an integral part of the official creed of Tibetan Buddhism, and the considerable cultural influence of Tibet spread Tibetan religious music to the nearby areas of Nepal, Sikkim, and Bhutan and, much later, to Mongolia.

Tibetan music. Tibetan religious music is the only Central Asian repertoire that has a long history of written notation. This notation, for liturgical chant, consists of neumes; *i.e.,* symbols representing melodic contour rather than precise pitch, similar to the earliest music writing of medieval Europe. Also distinctive is the metaphysical aspect of Tibetan Buddhist music, related to Indian philosophy. Each instrument of the monastery orchestra, as well as the drawn-out tones of chant, is believed to represent an externalized form of the *mantras,* or sounds inherent in the human body, accessible otherwise only through steadfast meditation. For the monks, such music is a basic aid to devotion and prayer. Musical styles vary somewhat among the sects of Tibetan Buddhism, but the basic approach and instruments are the same.

The monastery instruments typify the crossroads position of Tibet. Some, such as the large cymbals, stem from China, while others (the majority), such as the conch-shell trumpet and handbells, can be traced to Indian influence and are found as instruments of Buddhist worship as far away as Japan. Still other instruments, such as the large oboe and the ten-foot metal trumpet, are perhaps Near Eastern in origin. One wind instrument, the short trumpet made from a human leg bone, seems to be of purely local invention. Similarly, the structure of the music seems basically Tibetan. It is founded on a principle of greatly prolonged dense, deep sounds, such as unison long and short trumpets with oboe, or the seemingly endless bass chant of groups of monks, whose long, drawn-out notes are punctuated by sharp, extended bursts of percussion. Each monk is said to be able to sing two or even three notes simultaneously.

Festival dramas

Much of this music emerges from monasteries only at festival time, when the great *'cham* (dance) dramas, which may last several days, are performed for the public's entertainment and edification. These plays, which generally show the triumph of Buddhism over Bon, the earlier shamanistic religion of Tibet, may involve hundreds of musicians in the guise of masked dancers with drums, backed by a large temple orchestra. Other types of public music also abound, such as secular, perhaps Chinese-related historical plays with an alternation of dialogue and songs with orchestral accompaniment. There is also a strong tradition of folk dance, which may include songs sung by mixed antiphonal choirs (*i.e.,* two alternating groups of singers). Minstrels ply their trade along the caravan routes and play instruments perhaps more related to general Central Asian traditions than to the Indian and Chinese background of religious music.

The music of Nepal, Bhutan, and Sikkim. Little is presently known about the music of Nepal, Bhutan, and Sikkim. Minstrels play a major role in the musical life of Nepal, where, under the influence of Indian practice, musicians are classified according to caste, and each group is distinguished by specific instruments and repertoire. There appears to be a great deal of both Indian-related and indigenous folk music in the three Himalayan kingdoms. Varieties of lute, such as the long-necked *damyan* of Nepal and its Sikkimese relative, may be linked to a similar instrument of the Pamir Mountains; while a Sikkimese flute having an outside air duct (in contrast to the inside duct of a Western recorder) seems to be a unique instrument. The Sherpas of Nepal and other Tibetan-related populations of the Himalayas, along with the thousands of Tibetan refugees now living in the area, maintain the traditions of Tibetan Buddhist religious music. The *mani-rimdu* dance-drama of the Sherpas, a variant of *'cham,* is a good case in point.

THE STUDY OF CENTRAL ASIAN MUSIC

In the West, the study of Central Asian music has until recently been restricted largely to travellers' accounts and analyses of small samples of music. By far the bulk of collection and study of Central Asian music of Turkistan and Siberia lies in the domain of Soviet scholars, who instituted systematic fieldwork as early as the 1920s; this literature remains largely inaccessible to the non-Russian reader. Mongol music was the subject of sporadic but intensive fieldwork by Scandinavian researchers in the 1920s and early 1930s, so that some of the traditional music culture was documented before Mongol society underwent the changes brought by war and the advent of socialism. Tibetan music has attracted increasing attention since the late 1950s, when large numbers of Tibetan refugees poured into the Himalayan kingdoms and northern sectors of India, thus making Tibetan music more accessible to outside observation. Afghanistan has been the object of intensive musical investigation only since the mid-1960s. Thus, outside Soviet contributions, Central Asia remains a lightly researched although quite fertile area of musical investigation. (M.S.S.)

Performing arts: dance and theatre

The performing arts have played an important role in the spiritual and social life of Central Asia, where they evolved as didactic art forms within a religious context. Performance, therefore, occurs in conjunction with some religious or special event. Two main types of performance predominate throughout Tibet, Nepal, Sikkim, Bhutan, Mongolia, and parts of the Asian U.S.S.R.: those related to shamanism and those derived from Buddhism. The performing arts of the Turkic peoples of Afghanistan and Turkistan are different from these because of the influence of the Islāmic religion and are discussed in the article ISLĀMIC ARTS. The Hindu influence found in Nepalese theatre and dance is treated in the article SOUTH ASIAN ARTS.

Although primarily intended to serve the overt purposes of religion, dance and theatre in Central Asia are performing arts with covert aesthetic values. Vocal and physical expressions of appreciation by an audience attending a performance depend on the graceful and rhythmic execution of hand gestures, body movements, and footwork. Aesthetic values are best expressed in the elaborate and artistic costumes, masks, and makeup, coupled with effective though crude stage effects and props.

Use of music in dance and theatre

Musical instruments play an important role in the performance of Central Asian dance and theatre. Usually a drum, but in some cases a string instrument, is used by the shaman to induce the ecstatic trance during which he symbolically journeys to the heavens or to the netherworld when playing the role of a psychopomp, or conductor of souls. Performances of the Buddhist monastic dance, known as *'cham,* and the Buddhist morality plays, called *a-che-lha-mo* ("older sister goddess"), were accompanied by a variety of instruments, especially drums and horns. There were large and small drums, short horns with fingering holes, and long horns, particularly the *dung-chen* (great conch shell) made of brass and extending many feet. The *dung-chen* with a deep haunting wail accentuates the macabre that is so much a part of *'cham.* The Tibetan guitar *sgra-synan* (pleasant sound) is a stringed instrument used almost exclusively by Himalayan peoples for folk song and dance.

Perhaps because of the subjectivism of their religions, it was not the custom among the peoples of Central Asia to carry out objective studies of elements in their cultures, and, therefore, no indigenous evaluation of their dance and theatre is available. Although a few manuals for the performance of shamanic rituals, music, and the *'cham* do exist, as do scripts for the *a-che-lha-mo,* much of the history and traditional staging of these forms of theatre and dance was handed down by oral tradition.

SHAMANIC RITUAL

Altaic horse-sacrifice ceremony

Teachings that spirits are responsible for unexplainable phenomena, such as disease and death, and that these spirits can be controlled by an individual with special powers, such as a shaman, evolved in many primitive societies throughout the world, including those of the nomadic peoples of Central Asia. The roles of the shaman include oracle, healer, sacrificer, and psychopomp, and each role calls for the performance of specific rituals. The earliest form of theatre and dance in Central Asia, these rituals developed into an often complex genre of the performing arts. The horse-sacrifice ceremony among the Altaic peoples of east Central Asia, for example, embraces a full range of dramatic elements despite the fact that like all shamanic ritual it is essentially a one-man performance. The ceremony, which lasts two to three days, is one in which the shaman undertakes a journey to the heavens. After having set the stage, the shaman symbolically releases the soul of a real horse and then, astride a goose-shaped device, he chases the soul of the horse, all the while imitating the noises of the goose and the horse. Capturing the soul of the horse, the shaman, with help from the audience, then kills the real horse, and the flesh is prepared. The next evening, the shaman offers pieces of the horse meat to the spirits, and, amid loud drumming and chanting, symbolically goes to the heavens on the soul of the horse while ascending a notched pole. As he ascends through the higher and higher heavenly planes, the shaman communicates to the audience important information, such as predictions about the success of the coming harvest and about epidemics and misfortunes that threaten and how to avoid them through sacrifices. The ceremony is followed by merrymaking and drinking.

Shamanism maintains that the soul of one who dies a heroic or violent death ascends to the heavens, but that the soul of one who dies from disease, which is caused by an evil spirit, must go to the underworld. The part of a psychopomp, or conductor of souls to the netherworld, was, therefore, another role commonly played by the shaman. The shaman guides the soul to its destination while narrating details of its journey to his audience. In some cultures, such as that of the Lolo, or Yi, in the mountains of southwestern China, the souls of all the dead are led in this manner to the underworld; while in others, such as that of the Tungus, a subarctic forest people of eastern Siberia, the shaman is only called upon to act as psychopomp if the soul of the deceased continues to haunt his residence.

The shaman also serves as the repository of tribal folklore and beliefs. Through dance and dialogue, he instructs the audience in the traditional teachings of their ancestors, and by passing his knowledge and techniques down to his successor, those teachings remain intact for future generations.

Rituals for curing the sick, guiding the soul of the dead to the netherworld, invoking a deity, or visiting the heavens are performed by the shaman in a state of trance induced by frenetic dancing to the music of a drum or a string instrument. Elaborate, symbolic costumes and ritual objects that are used in the ceremony provide a dramatic and mystic spectacle. The expectations of the audience are directly connected with the purpose of the shamanic performance; but whether it is the hope that the patient would be cured or that the oracular communication be auspicious, those attending the shamanic performance do so with the expectation that the ritual will be an entertaining religious experience.

Tungus shaman, detail of an engraving from Witsen's *Noord en Oost Tartarye,* 1785.

By courtesy of the trustees of the British Museum; photograph, J.R. Freeman & Co. Ltd.

Regional variations in costumes

Before the introduction of Buddhism in shamanic Central Asia, there were no centres for the performing arts in the usual sense of the word. Each shaman performed his dramatic arts at his own residence or environs as the occasion demanded. He had his own ritual costumes and paraphernalia, which displayed regional variations, particularly in ornamentation. The representation of animals and birds is common, and metallic objects, which are thought to possess a soul and do not rust, are also important. For example, the costume of a Siberian Yakut shaman must have from 30 to 50 pounds (15 to 25 kilograms) of iron to be efficacious, while a Siberian Buryat shaman, except for

Paul Popper (Popperfoto) Ltd.

Oracle priest of Tibet in a trance.

an iron casque, or helmet, wears mostly furs. The metal ornaments represent such diverse things as the internal organs, bones, a woman's breasts, the Sun or the Moon; but the object common to all shamans is a metal mirror, in which the shaman can see the souls of the dead. Regardless of the variations in dress, the purpose and performance of the rituals remain essentially the same, whether carried out by a Buddhist monk among the Sherpas of Nepal, or by a true shaman among the Siberian Yakut.

The shamanic rituals of the steppe and desert peoples have analogies among the dramatic arts of the Himalayan kingdoms, where, because of the tolerance of local beliefs and rituals, many shamanic practices were adopted into Tibetan Buddhism. For example, the State Oracle of Tibet, a monk whose oracular powers were exercised on behalf of the government and the monastic system, was regarded as a high ranking ecclesiastic, yet his ritualistic performances were no different than those of shamanic mediums throughout Central Asia. The adaptation of the psychopompic role of the shaman into Tibetan Buddhism resulted in the recitation of the *Tibetan Book of the Dead* to an effigy of the corpse. This book describes in detail the frightening apparitions the deceased encounters day after day while in the forty-nine-day interval between death and rebirth, and its reading is analogous to the shaman's narration of his journey to the underworld.

BUDDHIST RITUAL

Buddhist monastic dance. The second major genre of the performing arts to develop in Central Asia was *'cham,* the ritualistic dance performed in Buddhist monasteries. The origins of *'cham* may well be an older form of shamanic ceremonial dance in Tibet, but centuries of evolution within a Buddhist dominated society led to the recasting of the roles and theme of the dance in keeping with Buddhist dogma. *'Cham,* which was introduced along with Tibetan Buddhism into Mongolia and parts of the southern Asian U.S.S.R. in the 16th century, became the main form of religious entertainment in eastern Central Asia.

Origins of *'cham*

The origins of *'cham* lie in Tibet's dim past, long before the introduction of Buddhism. Initially, it was performed as a ritual to drive out evil spirits and to appease the guardian spirits by means of human and animal sacrifices, thus assuring an auspicious and prosperous new year. According to Tibetan tradition, the ancient shamanic dance was adapted as a Buddhist one by Padmasambhava, the Indian tantric teacher who introduced Buddhism into Tibet in the 8th century AD. He is said to have interpreted the dance as symbolizing the victory of Buddhism over the shamanism of Tibet, and, since blood sacrifices are abhorrent to a Buddhist, these traditional elements were simulated by clever techniques using effigies and red-coloured substances.

Sectarianism developed in Tibetan Buddhism in the 11th century, primarily as a reaction to the unreformed teachings of Padmasambhava and his followers; nevertheless, each of the sects retained the monastic dance as part of their religious repertoire. The reformed Yellow Hat sect changed the time for its performance from the birthday of Padmasambhava to the end of the official year, which would coincide with the lunar month from the middle of January to the middle of February.

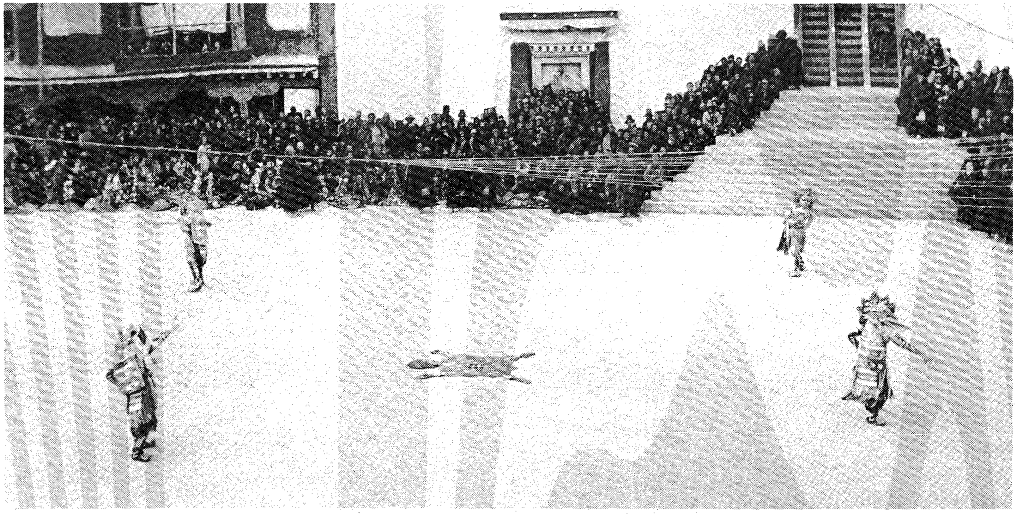
The acceptance and spread of Buddhism led to the eventual establishment of monastic communities throughout eastern Central Asia. These monasteries became fixed centres for the performance of *'cham.* Every monastery of adequate size and monastic population maintained its own masks, costumes, props, and musical instruments. In spite of regional and sectarian variations, the performance of *'cham* remains basically the same. The stage is set outdoors in the courtyard of the monastery called the *'cham-ra* (dance enclosure). With the exception of high lamas and members of the nobility who sit on special seats, the audience stands or sits around on the edge of the dance floor or ground. The musicians with horns and drums take their places, usually under a cloth canopy. Then, accompanied by music, the various dancers emerge from a building or from behind a stage curtain and perform. The first to appear are dancers wearing wide-brimmed black hats, topped with a simulated human skull. The costume of these dancers has led to the performance being referred to in some Western works as the Black-hat devil dance. Although of shamanic origin, the costume of the black-hat dancers is said by the Buddhists to represent the black disguise worn by Dpal-gyirdo-rje, a 9th-century Tibetan monk, who assassinated the fanatic anti-Buddhist king, Glang-dar-ma.

Black-hat dancers

The black-hat dancers are followed by a variety of performers, including those wearing monstrous masks representing a host of evil spirits that harass mankind, those costumed as skeletons and wearing skull masks, and those representing Indian teachers of Buddhism. There are also masked dancers representing the tutelary deities of Buddhism, and the most impressive of all is the Choskyi-rgyalpo (King of the Religion), who wears a mask fashioned after the head of a bull, which is emblematic of the aspect of the deity that vanquishes the Lord of the Dead. It is this dancer who dismembers an effigy of a corpse and scatters the parts in a simulation of the sacrificial and expulsional elements of the ancient shamanic dance rituals.

The dance is not all macabre. Comic relief is provided by a dancer wearing a mask with an expression of stupidity. This buffoon represents Ho-shang, the Chinese monk, who was defeated in an 8th-century debate on the merits of Indian versus Chinese Buddhism. Ho-shang is represented in the *'cham* of the Sherpas of Nepal by a dancer wearing a mask portraying a balding, bearded old man, called Mi-

(Left) Paul Popper (Popperfoto) Ltd., (right) EB Inc.



(Left) *'Cham* performance during the New Year celebrations at the Potala, Lhasa, Tibet.
(Right) *'Cham* dancer representing the Spirit of Barren Mountain, near Gangtok, Sikkim.

tshe-ring (Long-Life Man), who delights the audience by his farcical antics and pratfalls.

The whole of the *'cham* performance, which takes two or three days, is a visual presentation of the fear of demons and monstrous creatures and the way in which Buddhism serves to alleviate that fear. The audience is reassured that the good forces of religion have neutralized the evil powers of demonic spirits, and so the new year will be a prosperous one. This religious dance is performed on varying scales of grandeur in monasteries throughout the Buddhist cultures of Central Asia, but the most magnificent of all is its performance in Lhasa for the Dalai Lama, the ruler of Tibet.

Buddhist morality plays. The last performing-arts genre to develop in Central Asia was the Buddhist morality play, called *a-che-lha-mo.* The plays are based on the lives of legendary and historical figures, and through costume and masks the ethnic origin and ethical character of the players are revealed. Folktales, as well as historical and Buddhist canonical literature, are sources for the stories presented in *a-che-lha-mo.* Most plays are about mythical heroes who prove that Buddhism and its virtues conquer all evil in the end; but there are those that tell the story of historical personages.

Origins of *a-che-lha-mo*

Although traditions among the Central Asian peoples are vague about the development of shamanic rituals and *'cham,* they are clear about the origin of *a-che-lha-mo* and even point out the historical creators of the art. Some scholars regard the plays as derivatives of Indian theatre, but Tibetan tradition claims that the first performance of a morality play was produced by Thang-stong rgyal-po, a famous bridge builder of the 15th century.

One story tells of some Tibetans who were building a bridge and found that whatever they assembled during the day, demons dismantled at night. Thang-stong rgyal-po, a holy man well versed in the capricious ways of demons, advised the Tibetans to stage a play to divert the attention of the evil spirits. By this stratagem, they were able to complete the bridge. This story would seem to indicate that the primary purpose of the play was to entertain, but another story illustrates that Thang-stong rgyal-po realized that religious teachings would have a greater influence on the people if they were dramatized, so he developed the morality play to serve this purpose. Regardless of his intent, *a-che-lha-mo* evolved as didactic entertainment, and Thang-stong rgyal-po is regarded by the Tibetans as its patron saint.

From primitive beginnings, the morality plays developed into a popular performing art, complete with stylized costumes and masks, complex scenarios and effective staging. The scripts are written in a dialogue song style called *rnam-thar.* There are at least nine traditional plays in the Tibetan repertoire.

Mythical and historical plays

The most common type of *a-che-lha-mo* is the drama based on legend and mythology which often reflects a strong influence of Indian theatrical tradition. An example is the play *'Das-log Snang-sa.* The phrase *'das-log* means to return (*log*) from the beyond (*'das*) and is used in Tibetan to refer to anyone who was believed to be dead and then returns to life and relates all that was witnessed in the netherworld. *'Das-log Snang-sa* is about a virtuous woman named Snang-sa who was unjustly accused of adultery and beaten to death by her jealous sister-in-law. When Snang-sa was led before the fierce Lord of the Dead (*gshin-rje*), he found that she was pure of heart and mind, and he allowed her therefore to return to life. Once home again, however, her husband and relatives began to mistreat her, so she became a nun. The play ends with Snang-sa flying away from the convent roof and disappearing like a rainbow in the sky.

Another type of *a-che-lha-mo* is the plays based on the lives of Tibetan holy men or pious kings. *Rgya-bza' Bal-bza'* ("Chinese-wife, Nepalese-wife"), for example, tells the story of king Srong-btsan sgam-po (died AD 649) and his two Buddhist wives: Wen-ch'eng, a Chinese princess, and Bhrikuti, a princess from Nepal. These three historical figures are believed by faithful Tibetans to have been incarnations of Buddhist deities, and their story is very popular with the audience.

The performance of a play may take more than a day. Narrative recitation is used to set scenes, delineate character, and give continuity to the songs that tell the main story. Comic sketches or dances are performed between the acts. Each character in a play has a distinctive costume or mask. Usually Tibetan male characters and the heroes wear no mask, while virtuous female characters wear flat, teardrop-shaped masks, particularly the serpentine deities. Three-dimensional masks are worn by evil foreigners, demons, and witches, and masks that cover the entire head are used to portray animals. Like the *'cham,* the morality play is performed outdoors, with the audience, except for the specially seated lamas and nobles, standing or sitting on the ground around the performing area.

The morality plays are performed for profit by groups consisting of lay men and women, and some have Buddhist monks or nuns as members. These troupes are generally associated with a given locale, such as a village or a monastery; however, they often travel to give performances on special occasions, transporting their wardrobe and stage props with them. In time, four of the most popular troupes were required by the Tibetan government to perform plays at the summer palace of the Dalai Lama. These performances were obligatory as a kind of taxation. There are still several of the *a-che-lha-mo* troupes to be found among the Tibetan refugees in Nepal and India, where their performances are in popular demand.

Mongolian plays of Noyan Hutuqtu

The *a-che-lha-mo* did not spread from Tibet into other parts of Central Asia until the 19th century. According to tradition, the Monoglian lama Noyan Hutuqtu (1803–56) studied *a-che-lha-mo* as performed in the Kokonor (Tsinghai) region of northeastern Tibet and then introduced his own adaptations of it at Tulgatu-yin keyid, his own monastery near the village of Saynshand, in Mongolia. For the first time, in 1832, he produced a repertoire of four plays, and their performance required 17 full days. The four plays were based on the textual biographies of the historical Buddha, the Indian Buddhist Teacher Atīśa, the revered Tibetan hermit-poet Mi-la Ras-pa, and the fabulous "Moon-cuckoo" (*Saran-u Kökügen-ü Namtar*), a mystical story of a pious prince who became a cuckoo bird living in the forest.

Noyan Hutuqtu's productions differed from the usual performance of *a-che-lha-mo* in certain respects. His repertoire did not include any of the plays traditional in Tibet. His actors, unlike those of *a-che-lha-mo,* wore no masks; instead they painted their faces, a makeup technique associated with Chinese opera. Whereas *a-che-lha-mo* is usually performed outdoors in the Himalayan regions, Noyan Hutuqtu had a special theatre constructed near his monastery. It was a three-sided, two-storied, mud-brick building with two stage floors; thus, two scenes could be performed at the same time. The upper stage represented the sky, the lower one the earth. There were stage exits on both sides and trap doors in the floor.

The first Mongolian actors were called *schabi,* or disciples, of the lama Noyan Hutuqtu. These men and women formed a regular troupe and were invited all over Mongolia to perform.

FOLK DANCE

Among the peoples of Central Asia, folk dancing occurs as a form of entertainment at social occasions, such as festivals, weddings, and other celebrations, and private parties. Often impromptu, folk dances are sometimes performed without the accompaniment of musical instruments, and the performers rely on singing and footwork to maintain the rhythm.

Formalized folk dance does not appear to have evolved among nomadic peoples of the steppe and desert regions, but such dances did develop among the sedentary agriculturists, particularly in the Himalayan regions, where troupes of amateur performers were formed for local entertainment. Some dances were performed by a group of men and women forming a circle; in others, the dancers faced each other in lines. The dance steps and body movement were performed according to a stylized routine, and the rhythmic beat was accented by a measured stamp of the foot. (T.V.W.)

Visual arts

PREHISTORIC CULTURES

Paleolithic cultures. The earliest artifacts discovered in the Central Asian region were found in Siberia and western Turkistan and are from around the 13th millennium BC. During the millennia that followed, migrants entered the region from various directions, regardless of the geographical obstacles they encountered. As a result, some of their artifacts correspond with those produced at a similar stage of development in more western areas; some finds from the northeastern part of Soviet Turkistan, for example, are related to certain objects made in Iran and Mesopotamia, and those from northwestern Central Asia are linked to eastern and central Europe by means of the Volga River and of Kazakhstan.

The Paleolithic (Old Stone Age) sites of western Turkistan are mainly concentrated in the Lake Baikal area. A cave in the Baysuntau Range containing the body of a Neanderthal boy aged about nine had been so carefully prepared that it is evident that the people who made his grave believed in an afterlife. The site of Malta, 50 miles (80 kilometres) to the southeast of Irkutsk, and that of Buret, 80 miles (130 kilometres) to the north, are noted for their mammoth-tusk figurines of nude women. They resemble Paleolithic statuettes of Europe and the Near East and probably served as fertility symbols or as representations of the great goddess, whose cult was widespread. Some of these figurines depict elegant, slender women, others heavy, corpulent ones. Of five found at Buret, one is unusual in that it is of a clothed woman wearing a one-piece trouser suit with a hood attached to it comparable to those still worn by present-day Eskimos. In recent years Paleolithic sites have been discovered south of Samarkand, and rock paintings have been found at Zaraut Say (Zaraut Stream) in the Babatag Range, 50 miles east of Termez, and in the Shakty Caves in the Pamirs. Executed in red ochre, they depict hunting scenes. Those in the Shakty Cave are the older and include a man disguised as a bird and other men wearing skins and shooting at wild oxen with bows and arrows.

Fertility images and rock painting

The invention of the bow is ascribed to the 10th millennium BC, the Mesolithic Period. Man's artistic development during this period is attested by a pottery fragment of a most expressive woman's face dating from the 3rd millennium BC and recovered from the site of Vosnessenovka in western Siberia.

Neolithic and Metal Age cultures. Many Neolithic sites have been discovered in the Asian U.S.S.R., and the number of Bronze Age sites is even higher. The majority were found on the middle reaches of the Yenisey River, especially in the Minusinsk Basin, where metallurgy developed early. They testify to the existence of three main, basically successive, yet often overlapping cultures: the Afanasyevskaya, Andronovo, and Karasuk, so-called after the villages near which each culture was identified.

By courtesy of the Siberian Department, Academy of Science of the U.S.S.R., Novosibirsk

Pottery fragment with a woman's face from Vosnessenovka, Siberia, 3rd millennium BC. In the Siberian Department, Academy of Science of the U.S.S.R., Novosibirsk. 15 cm × 13 cm.

Afanasyevskaya burials

A cemetery to the southwest of Krasnoyarsk, on the slopes of the Afanasyevskaya Mountains, contained 80 burials dating from the 2nd millennium BC. The earlier ones were flat and marked by stone circles symbolizing the Sun god; the later ones took the form of barrows, or large mounds of earth, but were also encircled by similar stone slabs. The earlier graves contained elongated, spherical pottery vessels with pointed bases decorated with herringbone patterns. In the later graves this type of ware was superseded by flat-bottomed pots usually associated with sedentary pastoralist cultures. The graves also contained numerous stone and bone objects. Although copper objects were rare, they heralded the dawn of a new cultural period, the Metal Age.

Houses of Andronovo culture

The Andronovo culture succeeded the Afanasyevskaya in the 2nd and 1st millennia BC. Although found to the southwest of Krasnoyarsk, it is more frequently encountered in western Siberia and Kazakhstan. The settlement and cemetery of Alekseevskoe (present Tenlyk), some 400 miles (600 kilometres) south of Sverdlovsk, is especially important, because its earth houses were designed for permanent habitation. Their roofs rested on logs, and each dwelling had a central hearth used for heating purposes with side hearths intended for cooking. Bronze objects were numerous, and workshops existed for working copper. The metal probably came from mines in the Minusinsk Basin, Kazakhstan, and the western Altai Mountains, the latter having been worked as early as the 14th century BC.

Dating from about 1200 to about 70 BC—the dawn of the Iron and historical age—the Karasuk culture was located in the Minusinsk Basin, on the Yenisey River and on the upper reaches of the Ob River. Its creators must have been in touch with East Asia, for certain bronze objects, notably elbow-shaped knives, are related to those used between the 14th and 11th centuries BC in China during the Shang period. Stone pillars topped either with ram's heads, stylized animal forms, or human figures have also been discovered. Dzheytun, northwest of Ashkhabad, in what is called the Kyzylkum desert, is the oldest known agricultural settlement in Central Asia. It possessed a thriving Neolithic flint industry.

Pottery in the Ashkhabad region

Annau, six miles (10 kilometres) southeast of Ashkhabad, and Namazga-Tepe, situated in the same region and occupying an area of some 145 acres (60 hectares), are important Bronze Age sites. The pottery vessels recovered from Namazga-Tepe are decorated with painted plant and animal motifs showing affinities with contemporary pottery wares from the Near East. Figurines of dogs and sheep were numerous, and a model of a house has also been found. At Karatepe, also near Ashkhabad, an agricultural society produced fine pottery from the 3rd millennium BC, but it reached its fullest development in the 2nd millennium BC in a series of vessels decorated with particularly spirited animal designs.

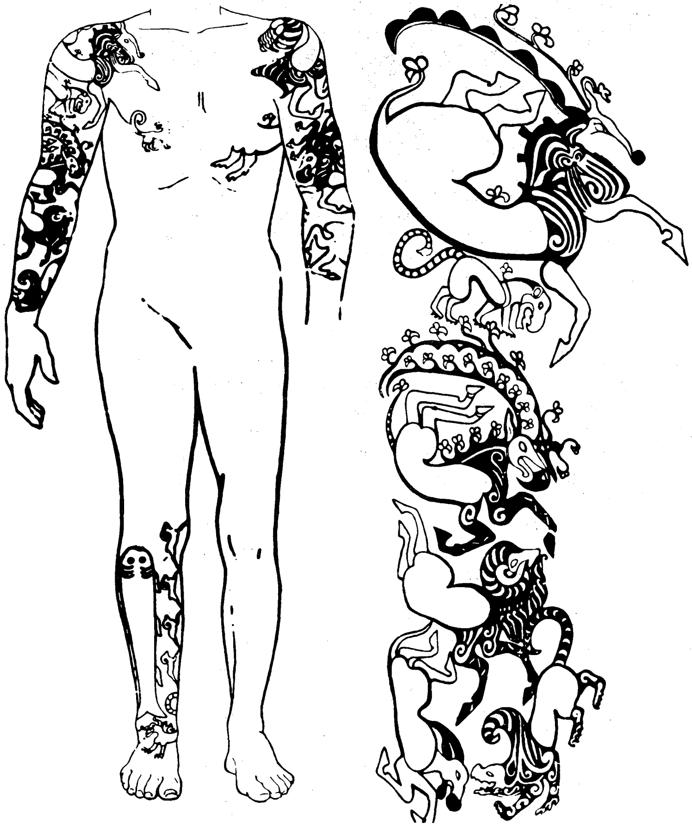
The main Neolithic and Bronze Age cultures produced several distinctive offshoots, which began to emerge by the early Iron Age. In Chorasmia, the Neolithic culture discovered at Dzhan-Bas-Kala is known as the Kelteminar and that of the Bronze Age, as represented by the Chorasmian level of Kokcha III, as the Tazabagyab. The Neolithic Hissar culture of southern Tadzhikistan spread across northeastern Central Asia into the Semirechiye, or foothills of the Tien Shan, while in Siberia the Bronze Age Karasuk culture was replaced in the 8th century BC by the Tagar culture. The latter endured until the 2nd century BC, producing an art of animal motifs related to that of the Scythians of southern European Russia.

The Bronze Age culture on Fergana's western border is associated with settlers living in large houses grouped to form settlements of considerable size. Some of the inhabitants worked in copper mines at the time when potters of the Chust Bronze Age culture of the Fergana Valley were producing fine-quality tableware, as well as cruder pottery articles. The best Chust pottery was very thin, covered with a red slip (liquid clay) and decorated after glazing with black triangular and scroll designs.

NOMADIC CULTURES

During the 1st millennium BC and the 1st centuries of the Christian Era, certain nomadic tribes affected the course of Central Asia's artistic history. Cyrus II the Great, the ancient Persian king who founded the Achaemenid Empire, was killed by the nomadic Massagetai when campaigning in eastern Iran in 530 BC. At the time, the Śaka tribe was pasturing its herds in the Pamirs, central Tien Shan, and in the Amu Darya Delta. Their gold belt buckles, jewelry, and harness decorations display sheep, griffins, and other animal designs that are similar in style to those used by the Scythians, a nomadic people living in the Kuban Basin of the Caucasus region and the western section of the Eurasian plain during the greater part of the 1st millennium BC. When considered together with objects of a like nature recovered from the frozen burial sites of the western Altai Mountains, it becomes evident that many of the Central Asian tribesmen commonly shared the traditions and culture that were once associated only with the Scythians.

By courtesy of the State Hermitage Museum, Leningrad



Tattooed designs on the male body found in Kurgan II, Pazyryk, 4th century BC. (Right) Detail from the right shoulder and right arm. In the Hermitage, Leningrad.

The Pazyryk burial site

Altaic tribes. Because of a freak climatic freeze, some of the Altaic burials, notably those of the 5th century BC at Pazyryk and neighbouring sites, such as Katanda, Shibe, and Tuekt, were isolated from external climatic variations by a protective layer of ice that conserved the organic substances buried in them. At Pazyryk these included the bodies of horses and an embalmed man whose body was covered with tattoos of Scythian animal motifs. The remarkable textiles recovered from the Pazyryk burials include the oldest woollen knotted-pile carpet known, the oldest embroidered Chinese silk, and two pieces of woven Persian fabric (Hermitage Museum, Leningrad). Red and ochre predominate in the carpet, the main design of which is of riders, stags, and griffins. Many of the Pazyryk felt hangings, saddlecloths, and cushions were covered with elaborate designs executed in appliqué feltwork, dyed furs, and embroidery. Of exceptional interest are those with animal and human figural compositions, the most notable of which are the repeat design of an investiture scene on a felt hanging and that of a semihuman, semibird creature on another (both in the Hermitage Museum, Leningrad). Clothing, whether of felt, leather, or fur, was also lavishly ornamented.

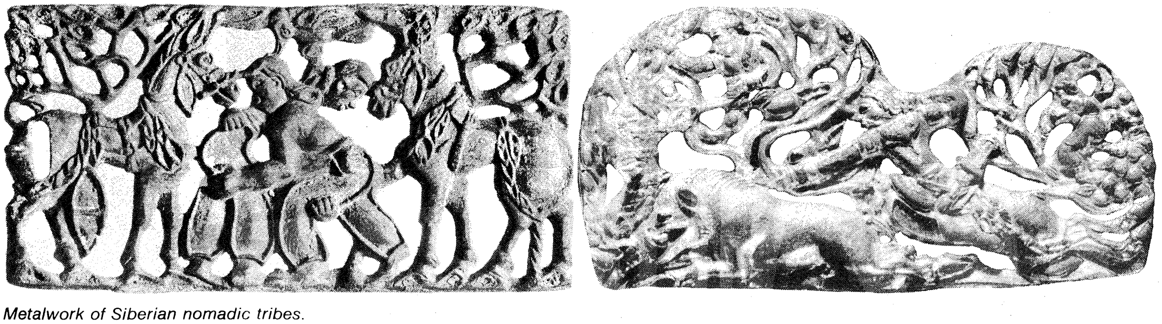
Horse reins either had animal designs cut out on them or were studded with wooden ones covered in gold foil. Their tail sheaths were ornamented, as were their headpieces and breastpieces. Some horses were provided with leather or felt masks made to resemble animals, with stag antlers or rams' horns often incorporated in them. Many of the trappings took the form of iron, bronze, and gilt wood animal motifs either applied or suspended from them; and bits had animal-shaped terminal ornaments. Altaic animals frequently display muscles delineated with dot and comma markings, a formal convention that may have derived from appliqué needlework. Although such markings are sometimes included in Assyrian, Achaemenid, and even Urartian animal representations of the ancient Near East, they seldom appear on those of purely Scythian origin. Roundels containing a dot serve the same purpose on the stag and other animal renderings executed by contemporary Śaka metalworkers. Animal processions of the Assyro-Achaemenid type also appealed to many Central Asian tribesmen and are featured in their arts.

Altaic decorative motifs

Certain geometric designs and sun symbols, such as the circle and rosette, recur at Pazyryk but are completely outnumbered by animal motifs. Such specifically Scythian features as zoomorphic junctures—*i.e.,* the addition of a part of one animal to the body of another—are rarer in the Altaic region than in southern Russia. The stag and its relatives, however, figure as prominently in Altaic as in Scythian art. Combat scenes between carnivores and herbivores that occur quite often in Scythian art are exceedingly numerous in Pazyryk work; but, whereas the Scythians show the victim passively accepting its fate, as on 5th-century BC gold triangular plaques from the so-called Seven Brothers burial in the Kuban, the Pazyryk beasts are locked in such bitter fights that the victim's hindquarters become inverted.

Siberian tribes. In the virtually contemporary metal-

By courtesy of (left) the Victoria and Albert Museum, London, and (right) the State Hermitage Museum, Leningrad



Metalwork of Siberian nomadic tribes.
(Left) Openwork bronze plaque from the Ordos region, showing two men fighting with their horses standing by; 3rd century BC. In the Victoria and Albert Museum, London. 6.5 cm × 12.8 cm. (Right) Gold B-shaped belt buckle from Siberia showing a mounted hunter chasing a wild boar; 1st century BC–1st century AD. In the Hermitage, Leningrad.

work of Siberian nomads, single animals of the cat family, such as panthers, carry the Altaic tendency of exaggeration further by twisting their bodies into a circle. In slightly later Siberian plaques, subtle openwork is used, a feature rarely present in Altaic or Scythian objects but frequently encountered in the more rounded versions of the animal style produced in the Ordos region of China, perhaps by Hunnish craftsmen, between the 4th century BC and the 2nd century AD. In the latter part of the 1st millennium BC, Siberian metalworkers adorned many of their gold and bronze plaques with artificial gems made of glass, as well as with jewelled inlays. They produced belt buckles shaped like the letter *B*. Two such gold prices (Hermitage Museum, Leningrad) are of particular interest because of their figural content. It has been suggested that they illustrate some ancient Central Asian epics, for one depicts a hunting scene and the other a warrior lying under a tree with his head resting on a woman's lap, while a servant holds their two horses. These subjects, possible forerunners of certain episodes in the *Shāh-nāmeh* ("Book of Kings," a work by the 10th century Persian poet Ferdowsī giving an account of ancient Iranian history), are thought to complement those on a series of openwork plaques, some of them of Ordos origin, on which either two dismounted riders are shown fighting while their horses stand passively on either side or two horses are seen locked in battle, pursuing their masters' quarrel (Hermitage Museum, Leningrad).

Mongolian Huns. In the 4th century BC the Huns started to migrate westward from the Ordos region. By the 3rd century BC, they had reached the Transbaikalia and had begun to enter Mongolia, which soon became the centre of their empire. Numerous mounds mark their progress. Those in the Zidzha Valley lie in the same latitude as the Pazyryk mounds and were subjected to similar conditions of freezing temperatures, which helped to preserve their contents. The richest of the excavated burial sites, however, are those of Noin Ula, to the north of Ulaanbaatar, on the Selenge River. Like those at Pazyryk, they included horse burials. The furnishings of one tomb were especially lavish. The prince for whom it was made must have been in contact with China, for his coffin was apparently made for him there, as were certain of his possessions buried with him (among them, a lacquer cup inscribed with the name of its Chinese maker and dated September 5, AD 13, now in the Hermitage Museum, Leningrad). His horse trappings (Hermitage Museum, Leningrad) are as elaborately decorated as many of those discovered at Pazyryk. His saddle was covered with leather threaded with black and red wool clipped to resemble velvet. The magnificent textiles in his tomb included a woven wool rug lined with thin leather (Hermitage Museum, Leningrad); the centre of the rug depicts a combat, of Scytho-Altaic character, between a griffin and an elk, executed in purple, brown, and white felt appliqué work. The animals' bodies are outlined in cord and embroidered. The design on another textile is embroidered in the form of a tiger skin with a head at each end. The animal's splayed-out body is formed of black and white embroidered stripes. Other textiles are of Greco-Bactrian and Parthian origin. In some of the Parthian fragments, Central Asian and Sāsānian Persian influences prevail over Hellenistic ones.

Chinese influences in the Noin Ula burials

Tashtyk tribe. On the Yenisey River, the Bronze Age Tagar culture was replaced by the Tashtyk culture, dating from the 1st to the 4th century AD. The physical appearance of the Tashtyk people has been preserved by a series of masks, some of them modelled; others, casts taken from the dead. They were painted with the features rendered in blue, red, and green against a yellow ground. Spirals disposed on the foreheads, temples, and cheeks of many of these masks probably represent tattoos. In many cases pearl necklaces worn by the women are also included. Although the animal motifs of the Tashtyks remained strongly Scytho-Altaic in style, the community was so much influenced by China that even its architecture was affected. Just south of Abakan, a large house made of beaten clay in the Chinese style has been discovered. Its roof had been covered with Chinese tiles, some of which carry inscriptions of the Han dynasty.

Painted masks

Fragment of a large rug from Noin Ula showing a griffin attacking an elk, 1st century AD onward. The woven woollen rug was mounted on fine leather, and then purple, brown, and white felt was appliquéd onto the rug over a cord binding, after which it was embroidered. In the Hermitage, Leningrad.

By courtesy of the State Hermitage Museum, Leningrad

PARTHIA

The Parthian Empire came into being in Khorāsān during the reign of Seleucus I, 358–281 BC, following the absorption by the Parthians of Parni (Dahae) tribesmen. The caravans traversing their territory brought them wealth and ideas from abroad. The figural art of the Hellenistic world made an especially strong impression on them. The finest Parthian objects come from Old Nisa, a town situated on the edge of the Karakoram Range, some 11 miles (18 kilometres) south of Ashkhabad in Tadzhikistan, close to the later town of New Nisa. Old Nisa was founded around 171 BC by Mithradates I to serve as a royal Parthian residence and necropolis, as well as the kingdom's capital. It contained several fine temples and an impressive palace built around a vast central hall, the roof of which was upheld by wooden supports set in stone bases—a practice followed in the town's larger houses. Lifesize clay statues of men and women stood between these supports. The royal treasuries contained many valuables, including silver and silver gilt statuettes of local Parthian deities and of Greek gods, bronze and iron weapons, burnished and painted pottery, glass, and cast bronze animals, such as griffins. The most significant of these treasures, however, is a series of ivory horn-shaped drinking vessels, or rhytons. Some are embellished

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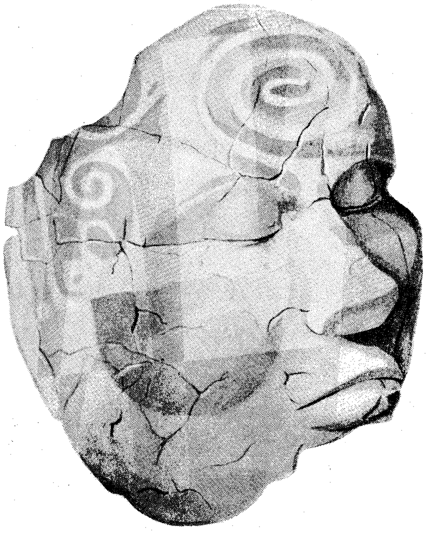
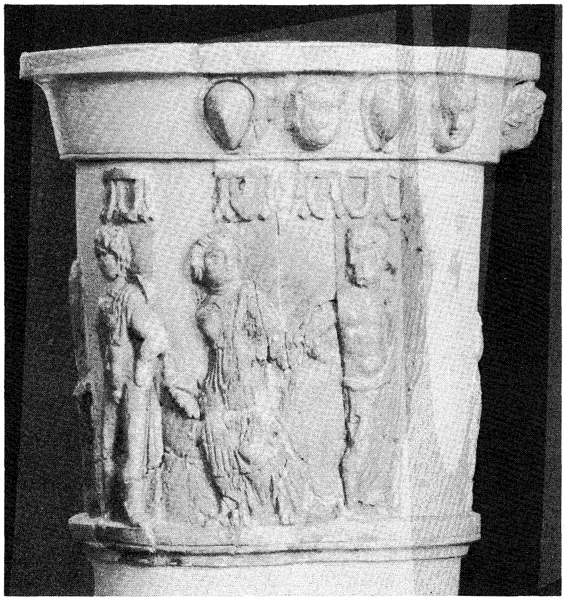


Death mask from a Tashtyk burial, 1st to 4th century AD.

with paste inlays and precious stones, others have a carved frieze or band encircling their open ends. One rhyton (Hermitage Museum, Leningrad) has a frieze of a procession that includes a Greek god. Conceived in the purest Hellenistic style, the frieze contrasts sharply with the rhyton's horned, lion–griffin-shaped terminal ornamentation, which is admirably modelled in the round, in accordance with the Scytho-Altaic tradition.

By courtesy of the State Hermitage Museum, Leningrad; photograph, Josephine Powell, Rome



Detail of an ivory rhyton from Nisa, 2nd century. The frieze shows a procession that includes a Greek god. In the Hermitage, Leningrad.

THE KINGDOMS OF WESTERN TURKISTAN AND AFGHANISTAN

Skill in irrigation, with the resulting expansion in agriculture, encouraged urbanism and the growth of states, changes that coincided with the rise of nomadism. While the nomadic cattle and horse breeders took over the steppelands, the culturally distinct states of Sogdiana (part of Uzbekistan and much of Tadzhikistan), Fergana (the greater part of Uzbekistan), Chorasmia (the Tashkent region), and Bactria (mainly Afghanistan) were established. At times independent, at other times reduced to vassaldom, the first three states were centred on rivers—Sogdiana around the Zeravshan and Kashkadarya, Fergana on the lower Syrdarya, and Chorasmia on the Amu Darya's basin. (The earliest references to these states are to be found in the Avesta, the principal scriptural work of the Zoroastrian religion, and in the inscription cut by order of the Persian king Darius I [reigned 522–486 BC] on the face of the rock of Bīsitūn in the Kermanshah province of Iran.) Bactria extended from the Syrdarya to the Hindu Kush (southern Tadzhikistan and Afghanistan) and is rich in unexplored mounds. Excavations at Balkh show that its first inhabitants settled there when others were doing so at Afrasiab (Samarkand) and Merv.

The political and economic changes that developed in the 4th century BC, following the Macedonian Greek king Alexander the Great's conquest of these states and their incorporation in the Seleucid Empire, and the conquests made, in turn, by the Parthians, Arabs, Turks, and Mongols are reflected in the regions' arts. The city of Alexandria-Kapisu Bagrām, founded by Alexander the Great, became the clearinghouse for India's western trade. India's religious beliefs, especially Buddhism, and the scriptural style that evolved in Gandhāra (an area situated between the Qondūz and Indus rivers in the Lower Kābul Valley of northwestern Pakistan) and Mathurā (in the Punjab region of northwestern India) followed along the trade routes and reached not only Bactria but, at times, also Kashmir, Tibet, China, and even the remote oasis towns of the Tarim Basin in Sinkiang. At the same time, Seleucid support resulted in the introduction of Greco-Roman art forms into Bactria, Kapisu, Taxila (Rāwalpindi), Gandhāra, Mathurā, and, after thirty years, even into Seistan.

Sogdiana. Sogdiana, with its capital of Afrasiab, was already noted for the sophistication and number of its towns when Alexander the Great conquered it in 328 BC and opened it up to Greek soldiers and administrators, and eventually to Roman traders. The Sogdians resented being governed by Alexander's successors, the Greek kings of the Seleucid dynasty. It is difficult to establish their relationship with their Seleucid suzerains and still more so with the later Kushans, but there is ample evidence to show that neither group of conquerors hindered the rise in both Sogdiana and Chorasmia of a local feudal nobility and class of rich farmers.

Pottery sculpture

A considerable amount of secular and religious pottery sculpture dating from the early Christian Era to the Arab invasion of the 8th century has been found at Afrasiab. The more interesting examples consist of statuettes of clothed women, some of them representing Zoroastrian deities such as Anahita. They have foreshortened bodies and large heads with a withdrawn expression on their faces and wear tiaras, hats, or hoods sewn to their cloaks. When the cloaks are sleeveless, they are worn over straight, long-sleeved robes instead of draped garments. All of the figures hold a piece of fruit, a symbol of fertility. Third-to-fourth-century statuettes from the fortified town of Tali Barsu, to the south of Samarkand, depict Syavush, the god of annual death and spring rebirth, as a musician. Statuettes of women flutists, riders, animals, and the Iranian semihuman-semianimal demi-god Shah Gopat have also been discovered there. In the 7th and 8th centuries, sculpture, whether in clay or alabaster, was highly developed at Pendzhikent, a site some 40 miles (60 kilometres) east of Samarkand, where Indian influence was often felt.

The earliest of Turkistan's mural paintings have been found in its eastern section. Those at Niy date from the 2nd century AD; those at Miran, from the 3rd. The inspiration in both stemmed from Rome, whereas Buddhism provided the impulses for the slightly later murals at Bāmiān and Kizil. In the eastern zone the paintings were designed as backgrounds to sculpture, and, as in western Turkistan, they were executed in tempera. Some very high quality murals recently discovered in western Turkistan are dated slightly later. The oldest ones, which are extremely fragmentary, are from the Varakhsha, a princely residence to the northeast of Bukhara, now lying in the desert; they date from the 3rd to the 4th century AD. Murals discovered at the beginning of the 20th century at Samarkand, which are almost contemporary with those at Varakhsha, have been lost. The importance of these murals is wholly eclipsed by the slightly later works discovered recently in Sogdiana, such as the 7th-century works at Varakhsha. Some of the rooms in the main apartments of the Varakhsha Palace (which consisted of several detached buildings) are decorated with high-relief alabaster stucco panels and carved woodwork, as well as with paintings. Benches are inserted into the walls of one room, the area above them divided into two registers, or horizontal rows, both painted red. In the upper register was a procession of animals, little of which survives, and, in the lower, splendidly attired hunters seated on elephants pursued spirited leopards and creatures of the griffin family.

Mural paintings at Varakhsha

Some 200 miles (300 kilometres) east of Samarkand, in a once fertile, now desert tract of land, the ruins of the great feudal castle of Mug survive. Among the objects excavated there was part of a wooden shield with the painted figure of a rider (Hermitage Museum, Leningrad), which foreshadows a type commonly found in Islāmic Persian book illumination. Mounted on a splendidly caparisoned horse, he wears a tunic of local cut and is equipped with a long sword, two daggers, two bows, and a quiver full of arrows. He is wasp waisted in the manner of figures depicted in murals of Varakhsha and Pendzhikent.

At Pendzhikent, a site close to Mug, and some 40 miles (60 kilometres) east of Samarkand, Sogdian architecture can be seen to advantage. The desert-engulfed city contained several large temples built of rectangular adobe

Temple architecture and decoration at Pendzhikent

bricks and blocks of beaten clay. The bricks were used for vaults and domes, while the flat sections of the roofs were made of rafters supported by wooden pillars or piers, some of which had been set in stone bases. Many of the more important houses were two storied. A square room measuring 26 by 26 feet (eight metres by eight metres) had served as a temple sanctuary. Although, in a series of rooms connected to it, some fragmentary religious paintings survived, the paintings in another temple are better preserved. They depict the death, the Sogdian burial rite, and the rebirth of a youthful Syavush. Over 50 figures of this vast composition survive, some representing Sogdian noblemen, some a group of Turks. A number of the Sogdians are seated crosslegged in the oriental manner and hold gold and silver vessels of Sāsānian shape in three fingers of one hand. The men's single, close fitting Sogdian tunics—elegant versions of the nomadic dress—resemble garments depicted in paintings of the Buddhist temples of Bāmiān and eastern Turkistan, notably at Kizil and Kuca. The style and, in some cases, the subject matter of these Sogdian scenes must have influenced the illuminators of such Islāmic Persian works as the *Shāh-nāmeh.* Another set of murals is unusual in that it is executed in high relief and then coloured. It shows human beings, sea monsters, and fish, with the waves of the sea rendered in lower relief than the figures. Yet another mural depicts a feast against a black background; a king and several priests sit crosslegged under a canopy; a woman harpist, some musicians and dice players, and a procession of elephants complete the scene. By placing light figures against dark or vivid backgrounds, Sogdian artists evolved a distinct form of perspective.

A study of the religious paintings shows that Central Asian Zoroastrianism retained elements from the earlier indigenous cult of the Sun and Moon. Some of the scenes in the secular works are linked in their subject matter (but not their style) to a small group of older Siberian gold and bronze *B*-shaped buckles and to the Siberian and Ordos plaques that are thought to illustrate local epics. Other secular scenes give full expression to Sogdian interest in the splendour of contemporary court life and prowess in hunting and warfare. The love of overall decoration and of animal motifs is as prevalent as in nomadic art. Details incorporated in Sogdian paintings proclaim the eclecticism of the society they depict and for which they were created. Sāsānian influence from Persia is seen in crowns trimmed with ribbons, veils, and bells; in the styling and trimming of hair and beards; and in many of their vessel shapes. The helmets worn by the warriors in the Pendzhikent libation scene resemble those depicted in the murals of eastern Turkistan. The clothes follow local fashions, and certain horse trappings display disks the shapes of which recall nomadic types.

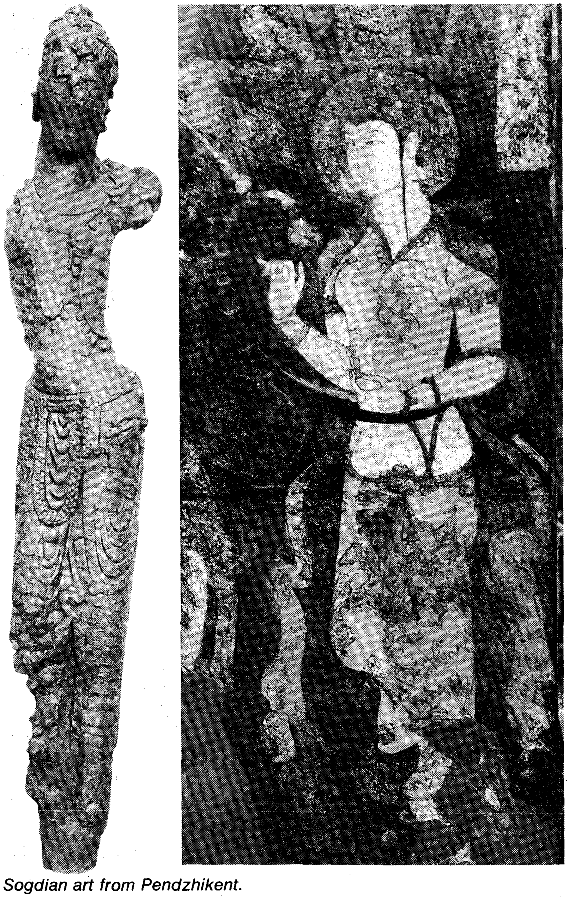
Sāsānian motifs in Sogdian textiles

Sogdian textiles are known to have been in great demand among their neighbours. Sāsānian motifs must have reached Sogdian weavers by way of imports from Persia, indirectly routed through Parthia, and also from Zoroastrians seeking protection in Sogdiana from Persian persecution. These motifs often figure both on surviving textiles and on those recorded in the paintings. The murals at Varakhsha, for example, include motifs taken from textiles; and a 5th-century mural from Balalyk Tepe displays the head of a tusked, boarlike animal set in a roundel that is almost identical to that on a Sāsānian fabric found at Astana in eastern Turkistan.

Between the 5th and 7th centuries, the Sogdians made dried-brick caskets shaped like rectangular rooms to contain ossuaries, or urns for the bones of the dead. The sides and lids of the ossuaries were decorated. The ornamentation on an ossuary from Bia Naiman (Hermitage Museum, Leningrad) has so many points in common with the decorations on a series of silver vessels that were, until recently, assigned to Bactria that the latter have come to be accepted as Sogdian. Several ewers have niches containing nude women rendered in a markedly Indian style, thereby recalling many a carved ivory plaque from Bagrām. Very similar niches adorn the Bia Naiman ossuary, but these contain crowned figures. In both cases the niches owe their form to Western influence, but those on the ossuary are formed of columns surmounted by capitals upholding pearl-studded arches, while on the ewers the Central Asian rosette replaces the capitals and the pearls.

Sculpture, both in relief and in the round, was widely produced in Sogdiana. Much of the earlier work takes the form of panels or friezes made of alabaster, stucco, and, occasionally, wood. Rosettes, roundels, disks, and vegetation provide the chief motifs. Audience chambers and large reception rooms often contained statues in the round. Even the statues attached to the wall had the appearance of being worked in the round. The earliest wooden caryatids, or columns in human form, are found in western Turkistan at Pendzhikent. The caryatids in the form of women have their hair elaborately dressed, and, although nude at the waist, they wear boleros, as well as close fitting, heavily trimmed skirts and splendid necklaces of Indian appearance. Once again, these figures recall those on Bagrām's ivory plaques and Buddhist statuettes of the 1st to 5th century.

By courtesy of the State Hermitage Museum, Leningrad; photograph, (left) Josephine Powell, Rome



Sogdian art from Pendzhikent.
(Left) Charred wooden caryatid, 7th century. In the Hermitage, Leningrad. (Right) A harpist, fragment of a mural in a private house, 7th–8th century. In the Hermitage, Leningrad.

Fergana and Chorasmia. Fergana produced much pottery of quality, but, as yet, there have been no finds of comparable importance to those in Sogdiana. Its arts appear to have paralleled the developments in the more prosperous, more heavily populated, and more highly urbanized state of Chorasmia (later Khwārezm). Chorasmia's defensive architecture was particularly notable. Its great citadels and palaces were enclosed by two lines of walls strengthened by massive towers that were fitted with lookout posts and firing slits and topped by archers' galleries. Chorasmian entrance gates were labyrinthine in plan. Many of these splendid buildings have disappeared beneath the desert's encroaching sands. Toprak kala, recently excavated, near Tashauz, is thought to have served not only as a citadel but perhaps as Chorasmia's capital until about the 7th century. Defended by stout walls, the palace of sun-dried bricks was equipped

Citadels and palaces

with three lookout towers. The ground floor of this two storied building acted as a foundation for the living rooms and storerooms above. Many of the rooms were adorned with sculpture: its most impressive room, the Hall of Kings, had niches fitted with grills ranged along the tops of its walls to hold statues of Chorasmia's rulers and notables; the Alabaster Hall contained many sculptures executed both in the round and in relief; a Hall of Victories contained statues of kings seated in the presence of a goddess of victory; statues of warriors carrying shields adorned the Warriors' Hall. All of the Chorasmian figural works are so lifelike that it is evident that portraiture had reached a high state of development by the 3rd and 4th centuries AD. Surviving decorations in the fortified manorhouse of Teshik Kala display the palmette, rosette, lotus, and ace-of-spade motifs that the Seljuqs later carried westward to Anatolia and beyond in the 11th and 12th centuries.

Bactria. The most Hellenized of these states in western Turkistan and Afghanistan was Bactria. Its fine coinage, for example, was distinctly Hellenistic in style, and Bactrian silversmiths were often influenced as much by Roman as Greek Hellenistic metalwork. Alexander the Great annexed Kābul to Bactria and founded Alexandria-Kapisu, a city astride the Indian caravan route, to serve as the province's capital. The multiracialism of Kapisu's population is reflected in the origins of the objects found there. Imports included articles from India, China, and the Greco-Roman world, especially from Syria. Artistic conventions characteristic of all these countries blended with the local Central Asian ones, with the Indian conventions predominating, to create Bactria's own distinctive style in sculpture, whether in alabaster, stone, ivory, or wood. Its mural paintings are wholly Buddhist in content, but they often contain features that link them to Fundukistan in India and the Sāsānian Persian world.

Importance of the decorative arts in Bactria

The decorative arts were highly developed in Bactria. Many of their sun-dried-brick houses were large enough to include several reception rooms, which contained many luxurious decorative objects.

Potters remained attached to animal forms derived from nomadic art. The large production of votive statuettes, especially representations of Anahita and Syavush, may be partly attributed to the belief that Zoroaster was born in Balkh. This tradition was also evident at Merv until the Arab invasion of Central Asia. The Bactrians mastered the technique of working metals at an early date. A 4th-century-BC lion-griffin (British Museum) in cast bronze is descended from a Scytho-Altaic prototype and so, too, is a pair of slightly earlier gold armlets (British Museum), originally embellished with inlay, from the Oxus Treasure. A series of silver dishes (Hermitage, Leningrad) from the end of the 1st millennium BC are, on the other hand, decorated with scenes drawn from the tragedies of the Greek dramatist Euripides and Greco-Roman mythology rendered in a markedly Hellenized style. Other silver dishes employ Indian motifs such as elephants. By the 8th century, these diverse ornamental motifs had fused, as on a silver-gilt bowl (Hermitage, Leningrad), dated between the 5th and 8th centuries AD, into a Kushān style that may well have provided the basis for Persia's later Rey figural pottery style.

By courtesy of the trustees of the British Museum

Bactrian lion-griffin, cast bronze, 4th century AD. In the British Museum. Height 24.9 cm.

KUSHĀN

The Kushāns replaced the Greeks in Bactria around 130 BC. They are thought to have been of Yüeh-chih stock with a strong admixture of Hephthalites, Śaka, and Tocharian. One branch of this group migrated to the Tarim Basin and founded a short-lived empire, while the other, under the name of Kushān, gained control of Central Asia. Capturing a section of the great trade route leading from India and China to the west, the Kushāns derived much of their revenue from the transit dues they exacted from the caravans crossing their territory, which often were carrying supplies of Chinese gold, silver, and nickel from the Tarim oasis towns to the Seleucid Persians. Around 106 BC, the first caravan to carry silk from China direct to Persia passed through territory that had belonged to the Seleucids but was now divided between the Kushāns and Parthians.

Kushān art reached its fullest development in the 2nd century AD, when the great king Kaniṣka is believed to have reigned. A magnificent, almost lifesize, now headless sculpture of Kaniṣka (Archaeological Museum, Mathurā) shows him wearing an elegant version of nomadic dress. His kingdom extended from Central Asia to include Gandhāra and Mathurā, where the Seleucids had so firmly established Hellenistic art that Western influence continued to maintain its hold even in the reign of the first members of India's Gupta dynasty. When Mahāyāna Buddhism reached Gandhāra during the 4th and 5th centuries AD, its sculptors turned to the Hellenistic world as a matter of course for a visual conception of Buddha and quickly evolved several Hellenistic versions. In the popular Apollo version Buddha is long faced, long nosed, and has wavy hair. This type survived into the 5th century and penetrated as far as Kashmir and Turkistan.

The Gandhāra and Mathurā styles of Buddhist sculpture

A school of religious sculpture equal in importance to that of Gandhāra developed almost simultaneously at Mathurā. Its earliest Buddhist images are virtually contemporary with the earliest ones produced in Gandhāra but, in Mathurā, Indian influences predominate. The portrayal of Buddha in the Mathurān style is softer yet more direct. The features are more Eastern; eyebrows extend in a continuous, sinous line; hair is straight; the earlobes are elongated; and an enigmatic smile replaces the withdrawn expression of the Hellenized Buddhas of Gandhāra. While the sculptors of Mathurā used a red sandstone, the Gandhārans worked in limestone or a local gray schist. They generally chose the latter for the small, uniform-sized panels with which they faced their *stūpas* and *vishanas*, carving them with scenes of Buddha's life. On the panels, the story unfolds from left to right, each scene being framed either within trees, leaves, or Corinthian columns sometimes linked by arches. These religious narratives often include furniture and details drawn from contemporary life. The figures, which use gestures of Indian origin to convey emotion, display racial characteristics that range from Indo-European to Mongolian. Many figures are presented in the frontal position favoured by Parthian artists, while others appear three-quarter face, as in Hellenistic art. Some wear Hellenistic robes and headdresses such as those worn in Palmyra, an ancient city in Syria. The Gandhāran style of sculpture was no longer produced after this area was invaded by the White Huns, in the 6th century.

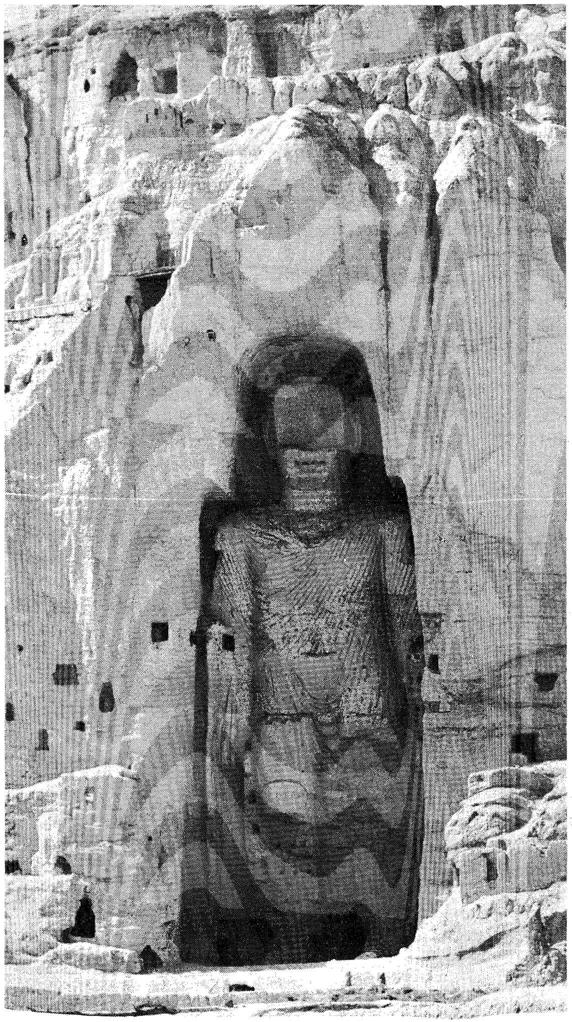
Central Afghanistan is rich in Kushān sites. Āteshkadehye Sorkh Kowtal, situated in the Qondūz Valley, close to the Kābul-Mazār-e Sharīf road, is dated by an inscription to the time of Kaniṣka's reign. The architecture of the region was very highly developed there. The town was protected by a double row of walls that ascended the hill on which it stood. The most impressive site within the wall was oc-

Mathurān and Gandhāran sculpture.
(Left) Indian-style colossal head of Buddha from Mathurā, red sandstone, Mathurā, 1st century AD. In the Archaeological Museum, Mathurā, India. Height 57 cm. (Centre) Birth of Buddha, limestone relief showing Hellenistic influence, Gandhāra, 1st–2nd century AD. In the Art Institute of Chicago. 27.4 cm × 50 cm. (Right) Apollo version head of Buddha from Gandhāra, gray schist, 2nd–3rd century AD. In the British Museum. Height 38.1 cm.

By courtesy of (left) the Archaeological Museum, Mathura, India; (centre) the Art Institute of Chicago, S.M. Nickerson Collection; and (right) the trustees of the British Museum

cupied by a dynastic fire temple, built to an Achaemenid plan in large blocks of well-dressed stone and approached by an imposing staircase. Within, columns topped by Corinthian capitals supported the roof. Numerous sculptures had originally adorned the interior, those worked with floral and animal motifs conforming to the Gandhāran tradition, while figural works followed Scytho-Parthian and, to some extent, Hellenistic traditions.

By courtesy of the Afghan Information Bureau, London



Colossal Buddha carved from the rock at Bāmiān, Afghanistan, 5th century. Height 51.8 m.

Buddhist art at Bāmiān

The Buddhist art of central Afghanistan is admirably represented at Bāmiān, where Mani, the Iranian founder of the Manichaean religion, probably lived and encouraged the growth of a religious pictorial art in the 3rd century AD. At both the eastern and western approaches to Bāmiān, a huge statue of Buddha as ruler of the world was cut into the face of the rock. The smaller statue measures 120 feet (about 40 metres) and dates from soon after the town's foundation in the 4th century AD; the other measures 170 feet (50 metres) and dates from the 5th century. In their commanding monumentality, both reflect the influence of the Mathurān image of King Kaniṣka and the portrait sculpture of Sāsānian kings and Parthian notables. Traces of painting show Sāsānian and Indian influences in the rock-hewn niche behind the earlier Buddha.

Regardless of Manichaean influence, Sāsānian elements prevailed at Bāmiān between the 4th and 6th centuries. At Dūktar-e Nowshirvān, near Bāmiān, a 4th-century painting of a Sāsānian king flanked by attendants survives. The murals in Bāmiān's 5th-century temple of Kakrak include one of a deified king of Sāsānian appearance, while others display the figure of Buddha set within a circle and wearing a costume of the Sāsānian type. Sāsānian motifs of paired birds and griffins placed in medallions or pearl circlets are common. In the murals at Imgur-Enlil, Buddha wears a close-fitting tunic resembling that worn by the Sāsānian king depicted on the rock carvings of Tāq-e Bostān. The traces of Hellenism, which are also evident in these wall paintings, began to disappear by the 5th century, when Sāsānian influence gradually gave way to the Gupta style of India.

Stemming from Gupta art is the practice adopted at Bāmiān between the 5th and 6th centuries of painting in the dome of a sanctuary a Buddha within a circle or hexagon. Gradually, these circles and hexagons became symbols of the heavenly Buddha. Many developed into rosettes and eight pointed stars—motifs that were retained in the 10th and 11th centuries by the Islāmic Seljuqs, who carried them to Persia and Asia Minor. As Gupta influence increased, sculpture gained in importance. A new style had evolved by the 8th to 9th centuries, but it did not penetrate into western Turkistan, where the Arab conquerors religiously opposed figural art. In the 9th century many Buddhists left Kashgaria, and Islām gained ground. Figural sculpture was forced underground and was primarily produced by secret shamanistic cults of an indigenous Central Asian origin. Although figural art was never to flourish in western Turkistan as gloriously as it had prior to the Arab invasion, there was a revival under the Mongols in the form of book illuminations.

EASTERN TURKISTAN

Uighurs. The figural arts found new patrons in eastern Turkistan among the Turkic Uighurs, who while living in

Buddha in a mandala preaching, a painting from the dome of a cave shrine at Bāmiān, Afghanistan, 5th–6th century.

Josephine Powell, Rome

Portraiture of the Turkic Uighurs

T'ang dynasty China had been influenced by Manichaean figurative art. The overthrow in China in 846 AD of Buddhism by official Confucianism forced the Buddhist Uighurs to migrate to eastern Turkistan. Gradually, they gained control over the Tien Shan region, Turfan, and the northeastern section of the Tarim Basin. The Turkic Uighurs especially favoured portraiture. In the 7th and 8th centuries Uighur artists already had acquired great proficiency in rendering likenesses in a style heavily influenced by Chinese portraiture of the T'ang period. These portraits were painted on silk and were frequently inscribed with the sitter's name.

Khitans. The figural style is believed to have been transmitted to the Mongols by the Khitans when the latter were living on the middle reaches of the Yenisey. The wealth of the Khitan princes is reflected in the furnishings of burial mounds discovered at Kopeni, some 200 miles (300 kilometres) to the south of Krasnoyarsk. Dating from the 7th to 8th centuries, these mounds were similar in type to those constructed by the nomads of the 1st millennium BC. One of the richest graves contained four gold jugs set on a silver dish and a number of gold, silver, and bronze ornaments (Hermitage Museum, Leningrad). Two of the jugs, although undecorated, carry Orkhon inscriptions on their bases. Two others are covered with delicate relief representations of birds and fish surrounded by flowers and vegetation, executed in a style influenced by Islāmic art. A Scytho-Altaic hunting motif of riders pursuing a tiger, a deer, and a panther appears on a bronze ornamental object.

Turkic tribes had been concentrating their numbers in Central Asia from about the 5th century AD. In the 6th century the Kul Tepe and Bilge Khan tribes established a state of their own in the Orkhon Valley. The inscriptions that they carved on the valley's rocks are of considerable historical importance. In the 7th century the Turkic Oğuz people were so numerous that they constituted 24 tribes. The Sāmānids, Ghaznavids, Ghūrids, and Seljuqs were of Oğuz extraction.

Türbe construction

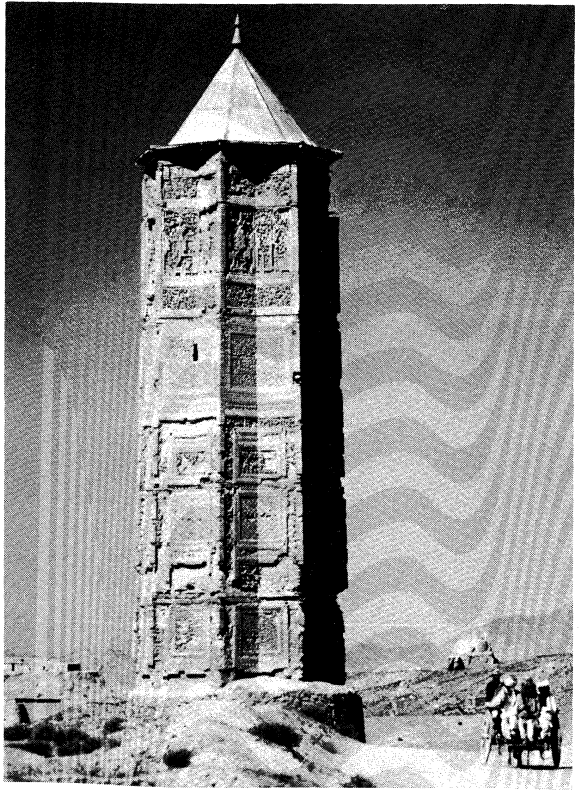
Sāmānids. The Sāmānids centred their kingdom in Khorāsān. In the 9th century, under the leadership of Esmā'il, they ruled over Transoxiana and eastern Persia from their capital of Bukhara. Esmā'il's *türbe,* or mausoleum, the oldest Islāmic monument surviving in Bukhara, reproduces the form of the Zoroastrian *chanar taq,* or fire temple. In Sāmānid and Seljuqid hands, the *türbe* generally took the form of a small circular or octagonal building, roofed with a turret shaped like the point of a pencil. Mounted on a solid or single vaulted substructure, its single chamber had a domed ceiling and a mihrab, or niche indicating the direction of Mecca. In the more elaborate *türbes*, the single door was framed with bands of geometric decoration, and the turret was sometimes ribbed.

Ghaznavids and Ghūrids. Alp Tigin, a slave of Turkic origin at the Sāmānid court, escaped in 962 AD to Kābul, where he rapidly gained control of the town. He transferred his headquarters to Ghazna in central Afghanistan and established his dynasty there. Few Ghaznavid works of art have survived, but the admirably proportioned and decorated mortuary towers at Ghazna are architectural achievements of great splendour. Still finer is the minaret of Jām, a Ghūrid structure of the 11th century. Standing alone in a desolate region, it escaped discovery until 1957. It is conjectured that the minaret may mark the position of the lost Ghūrid capital of Fīrūzkūh.

Seljuqs. The art of the Seljuqs, who founded kingdoms in Persia, eastern Byzantium, Syria, and Iraq, eclipsed that of the Sāmānids, Ghūrids, and Ghaznavids. They were great architectural patrons and constructed numerous mosques, madrasah (Islāmic religious schools), hospitals, orphanages, baths, caravanserais, bridges, and *türbes* notable for their decorative masonry, elaborately ornamented portals, and use of Kūfic script as an architectural decorative device. The Seljuqs also attained a high standard in their decorative arts, especially metalwork, wood carving, and pottery. The Mongols, who terminated the Seljuq period, adopted certain Seljuqid artistic conventions, particularly the use of ornamented portals and glazed-tile paneling.

Mongols. Genghis Khan (died 1227), the renowned Mongol conqueror, sacked and destroyed Bukhara in 1224, sparing only the 12th-century Kalyan tower, which was used for throwing criminals to their death. The 14th-century Turkic conqueror Timur, however, endowed Samarkand with new glory by building a series of religious monuments widely renowned for their splendour and decorative use of glazed tiles. In the 16th century, Bābur, prince of Fergana, coveted Samarkand. Failing to capture it, he chose Kābul as his headquarters for his conquest of India. His tomb there (he died in 1540) is the only visible testimony to the years he spent in the city.

Josephine Powell, Rome



One of the two mortuary towers at Ghazna, Afghanistan, 10th–11th century.

Emirate of Bukhara. Although fine-quality pottery decorated with animal, bird, and figural designs was being made in New Nisa in the 15th century, the artistic revival of the Mongol period that Timur had launched in western Turkistan had died out by the 16th century, when the emirate of Bukhara, incorporating much of Sogdiana, was established. Except for gold-thread embroidery and carpet making, in most of Central Asia the visual arts largely stagnated. In Mongolia the conversion of the Buryats to Lamaism in the 18th century brought into their tradition of ornamentation such Tibetan motifs as the lotus, dragon, and lion.

Revival of traditional architecture and decoration

Russian–Soviet period. In 1882 the emirate of Bukhara was incorporated as a Russian state. This political act had little cultural effect, and European art remained unknown to Central Asians. Traditional indigenous architecture of baked or unbaked brick construction was revived in the 18th century. Carved doors and screens were again produced. Old styles of Islāmic script were combined with arabesques to adorn metalwork. Zoomorphic junctures persisted in the animal designs created by metalworkers and potters alike, although the ornaments worn by nomadic women had become so stylized as to have lost all resemblance to the ancient animal motifs from which they were descended. Openwork remained a feature of much of the jewelry, notably of the necklaces formed of small openwork plaques linked by rings or chains.

Following the Russian Revolution, a new phase of art began in the Soviet-controlled regions of Central Asia. Although the Soviet authorities took steps to maintain the existing carpet and textile industries, they encouraged the inclusion of genre scenes and native animals and vegetation. They also founded schools to train artists in the traditions of European art. Pictorial arts are naturalistic in style, conforming with the principles of social realism as defined by the Soviet authorities. The first Buryat-Mongolian Turkmen painter to achieve distinction in this style was Tsyrenzhap Sampilov.

By courtesy of the University of Alaska Museum, Fairbanks

Walrus ivory carvings from the Arctic.
(Left) Ipiutak rake or comb showing the head of a bear (or seal), found near Spencer Point, near the western tip of the Seward Peninsula in northeastern Alaska, 1st–6th century AD. Length 26 cm. (Right) The Okvik Madonna, found on an island off the southeastern coast of St. Lawrence Island in the Bering Sea, 1st–5th century AD. Height 17.1 cm. In the University of Alaska Museum, Fairbanks.

ARCTIC REGIONS

In the arctic zone of Central Asia, the prehistoric age extends from the 3rd millennium BC to the arrival of Europeans around AD 1800. Knowledge of the region's arts is still very limited, for it is wholly dependent upon the sculptures produced by Eskimos living on the shores and in the hinterland of Siberia and the Bering Strait. These sculptures are mostly in walrus tusk, though wood and reindeer horn examples also exist. The majority are small in size and worked in the round to form terminal ornaments for utilitarian or ceremonial objects or statuettes. The latter are not provided with bases and thus must have been designed to be carried about. Many of the implements are decorated with incised patterns formed chiefly of lines and dots. As in all early arts, the statuettes and terminal ornaments are largely concerned with hunting or the magical practices of shamanism. The earliest and finest statuettes of which there is knowledge are assigned to the Okvik culture, which some scholars date to the pre-Christian era, but which others assign to its early centuries. Okvik art is concerned primarily with the representation of the human figure, differing in that respect from the contemporary or slightly later Old Bering Sea culture, where interest largely centres on animals, such as reindeer, elks, bears, and seals.

Okvik sculpture

Works of the Okvik and later Arctic schools often depict women, sometimes in the nude, sometimes clothed. The nude figures seldom include more of the arms than shoulder stumps. Their bodies are short and flat, their heads large, pear-shaped, and carefully worked, as are hands when included. The faces are carved and are sometimes incised with lines, probably denoting tattooing. The so-called Okvik Madonna (University of Alaska Museum, Fairbanks) is perhaps the most expressive of these statuettes.

Some Okvik animal designs are particularly interesting because of certain stylistic details that point to a relationship with works of the Scytho-Siberian school. Reindeer are so frequently depicted that the discovery at Pazyryk of a horse's mask in the form of a reindeer's head led to the suggestion that the mask was a survival from a reindeer cult acquired by the Altaians from a northern people such as the Eskimo. That theory has been discounted, yet some Okvik works are undoubtedly related to certain slightly older examples of Siberian metalwork. Thus, the heads of some terminal ornaments bear a close resemblance to those of certain Siberian works. The lozenge-shaped muscles that appear on Eskimo carvings amid lines intended to portray the animal's skeleton are very similar to those of the Pazyryk dot and comma markings. In late- or post-Okvik times certain specifically Eskimo objects, such as masks, were decorated with stylized animal heads executed in relief and accompanied by bosses that recall the Altaic, especially those that reflect Chinese influence. Compositions such as that on an unidentifiable object (possibly a rake or comb) in the University of Alaska Museum, which includes as its central motif the head of an animal resembling a bear or a seal, display a marked affinity with west Siberian ones.

Climatic changes in the 17th century led to contacts with the outer world in the 19th century and brought the traditional Eskimo school of sculpture to an abrupt end. When, toward the end of the century, art started to revive, it did so under European influence, eventually developing a greater concern for aesthetic than religious considerations. The new style retained much of the directness of approach and formal conventions of the traditional style, but, in addition, there was a greater emphasis on naturalism. Group scenes, too, became popular, as did animal and bird compositions. There has been an extensive production of small sculptures, chiefly of fish, bird, or animal forms, in the 20th century. (T.T.R.)

HIMALAYAN CULTURES OF NEPAL AND TIBET

Nepal. The art of Nepal is centred in the Kāthmāndu Valley, in an area of less than 250 square miles (650 square kilometres). The artists are Newars, or Mongoloids, different ethnically from, though partly intermingled with, the peoples of India, whose art they made their own—whether its themes were Hindu or Buddhist.

A rider on an elephant attacked by two leopards, Sogdianan fresco from Varaksha, possibly 7th–8th century AD. In the Hermitage, Leningrad. Height about 1.2 m.

Art of Turkistan and Afghanistan

Griffin armlet from the Oxus Treasure, Achaemenid period (559–330 BC). In the British Museum.

A female donor, fragment of a wall painting from Bezeklik in Eastern Turkistan. In the Museum für Indische Kunst, Staatliche Museen Preussischer Kulturbesitz, West Berlin.

Yakshi (River Goddess), ivory carving from Bagrām, Afghanistan. In the Kābul Museum, Afghanistan. Height 46 cm.

Plate 1: By courtesy of (top) the Hermitage, Leningrad, (bottom left) the Museum fur Indische Kunst, Staatliche Museen, Berlin, (centre) the trustees of the British Museum, (bottom right) the Kabul Museum, Afghanistan; photographs, (top and bottom right) Josephine Powell, Rome, (centre) EB Inc.

Art of Tibet and Nepal

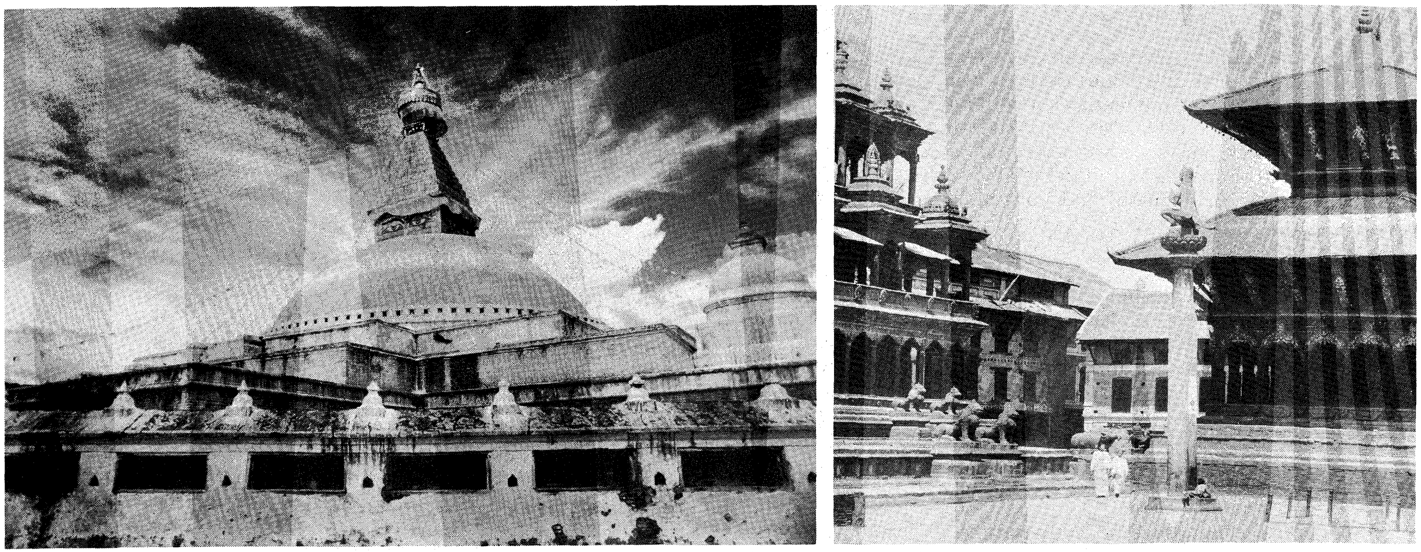
Taleju Bhavani Temple door facade, Bhaktapur, Nepal, 17th century.

Reliquary casket of gilt bronze from the monastery dPal-l'or c'os-de in Gyangtse, southern Tibet, 16th century. In the Rijksmuseum voor Volkenkunde, Leiden, The Netherlands. Height 35 cm.

Hayagrīva, tanka from Tibet, early 19th century. In the Rijksmuseum voor Volkenkunde, Leiden, The Netherlands.

Ratnasambhava, the Dhyāni Buddha of the south, surrounded by the eight *mahābodhisattvas,* Nepalese painting.

Plate 2: By courtesy of (top right and bottom right) the Rijksmuseum voor Volkenkunde, Leiden, The Netherlands, (top left) from M. Singh, *Himalayan Art* (UNESCO, 1968); photographs, (top left) Madanjeet Singh, (bottom left) Holle Verlag



Nepalese architecture.
(Left) Bodnath *stūpa*, near Kāthmāndu, *c.* 13th century. (Right) Temples on Durbar Square, Lalitpur.

(Left) Larry Reynolds; (right) Josephine Powell, Rome

Architecture. There is only one Nepalese architectural style, varied according to its function as private dwelling, palace, Buddhist monastery, or Buddhist or Hindu temple. The style is the protracted local flowering of an Indian architectural tradition—of brick and wood architecture with tiered, sloping roofs—other varieties of which are found in the western Himalayas and in Kerala in the southwest.

Essentially, there are two kinds of Nepalese Buddhist shrine, or *stūpa* (also called *caitya*): the large *stūpa* and the small, monolithic *stūpa.* Characteristic of the large *stūpa* like the one at Bodnath is the low base from which it rises and its crowning dome-shape. The small *stūpa* was generally set in the courtyard of a Buddhist monastery. The extant monasteries, none of which dates earlier than the 14th century, are consistent in their plans and structures. A central courtyard flanked by residential buildings is entered through a gate with a richly carved tympanum (torana) and porch. Opposite the gate and in the centre of the courtyard is the main building, the *stūpa;* with its one- to three-tiered roof, it rises higher than the buildings that surround it and forms the square of the courtyard.

Buddhist monasteries

Most Hindu temples are freestanding. The more ancient temples have two superimposed roofs; the later ones are five-roofed temples, given further height by tiered brick socles, or bases. On each story of the towerlike structure, wooden beams and struts (a structural piece designed to resist pressure in the direction of its length) support a widely projected slanting roof, the struts ascending diagonally from the central structure to the edge of the tiled roof. The majestically tapered, ascending profile of the structure, with its strong contrast of light playing on the roofs and masses of shade looming below, is peculiar to Nepal. Rich in textures and colours, the temples are embellished with carved and painted struts, carved doorframes and window frames, and embossed gilded copper sheets. Like the pantheon on the stone temples of India, the pantheon of Nepal is laid out mainly on the exterior of the temple—in contrast to Tibet, where it is displayed on the interior of the temple.

Sculpture and painting. Combinations of Hindu and Buddhist iconography came about easily, though there is something facile about them, a smoothness found also in the form of the Nepalese images, which lack the surging dynamism of Indian form. Characteristic of the Nepalese transformation of Indian styles is a loss of depth but a gain in grace. Suavity of line, temperance of modelling, tonal clarity of vivid, contrasting colours raise Nepalese works far above the merely derivative. An indigenous physiognomy, too, modifies the physical formulas for sculpture laid down in India.

While Nepalese sculpture is known to exist from the 2nd century BC (terra-cotta plaques, a stone *bodhisattva,* and a Buddha image), it was in the 5th to the 7th centuries BC that stone sculpture in Nepal came into its own. Vishnu Vikrānta (the three strides of Vishnu), dated AD 467, and 6th-century panels illustrating the Kumārasambhava ("Birth of the War-God," an epic by the 5th-century Indian poet and dramatist Kālidāsa) are masterworks of narrative relief and dramatic mythical composition. On the more intimate level of daily life, sculpture takes the form of the many fountains that adorn watering places (*pranali*) of Nepal. Water spouts forth from *makara* (Hindu water monster with the body of a crocodile and the head of an elephant) snouts sheathed in gilt copper into reservoirs laid out with architectural dignity. As far as present knowledge goes, Newari sculpture was dominated from the 8th century into the 18th by gilt-copper images. In their glowing splendour, the gilt, sometimes jewel-encrusted images embody the Buddhist quality of compassion that leads to enlightenment.

Painting in Nepal is known from the 11th century on palm leaves and wooden bookcovers of manuscripts, some of them hardly distinguishable, at first, from the Bengali prototypes. The Nepalese style, less nervous, more conscious of the beautiful line and clear, compartmental order of the surface, is fully developed in scrolls, or *prabhas* (most of them, vertical), on cotton known from the 13th century. These scrolls are of two kinds: one consists of ar-

By courtesy of the Prince of Wales Museum of Western India, Bombay

Detail of the Rathayātrā Scroll, Nepalese, 1617. In the Prince of Wales Museum of Western India, Bombay. Height 64.2 cm, width (entire scroll) 252.7 cm.

rays of religious images with a large figure of the main deity in their midst; the other consists of a *maṇḍala,* the Hindu and Buddhist symbol of the universe—a circle enclosing a square with the deities disposed within. Narrative panels or sections in the margins of both types of scroll soften the rigour of the composition. While this Nepalese hieratic, or sacerdotal, style was at its peak, a narrative style developed in manuscript illuminations such as the *Hitopadeśa* (1594; Kāthmāndu) and horizontal scroll paintings such as the Rathayātrā Scroll (1617; Prince of Wales Museum of Western Bombay). Its planar intricacies reveal a new and vital aspect of Nepalese painting, an immediacy of emotion and action of its protagonists, the figures of which are placed on an opaque, velvety ground. The colours of these book illustrations and scrolls retain the strength and depth of those of the hieratic scrolls, which continued to be painted into the 17th century. The influence of the more realistic Indian, Rajasthani paintings, from the latter part of the 17th century, finally overwhelmed the hieratic style. Its disappearance was further hastened by a wave of Chinese-influenced Tibetan painting. (S.Kh.)

Tibet. Tibetan art comprises ancient pre-Buddhist decorative and domestic crafts and the all-pervading religious art that was gradually introduced from the 8th century onward from surrounding Buddhist countries and developed subsequently as recognizably distinct Tibetan imagery, sculpture, and decorative architectural motifs. In all its forms Tibetan art has remained subservient to special lay or religious intentions and has never become an art pursued for aesthetic ends alone. The religious art is primarily didactic and symbolic; the lay art, decorative. Therefore, while lay art may be easily appreciated, to understand the significance of the religious art requires knowledge of Tibetan religion and religious symbolism. Since the destruction of Tibetan cultural traditions by Chinese-trained Communists from 1959 onward, a greater interest has arisen in the West in the surviving Tibetan *objets d'art* preserved in museums and private collections.

Up to the 9th century AD, Tibet was open to cultural influence from Central Asia, especially Khotān, and from China. For two centuries, up to the collapse of the old Tibetan kingdom in 842, the Tibetans controlled the whole Takla Makan and the important trade routes from the Middle East to China. Stone carving and metalwork were certainly practiced in the pre-Buddhist period, and Persian, Indian, and Chinese influences, all received through Central Asia, have been noted.

The introduction of Buddhism from the 8th century onward led to the arrival in Tibet of Buddhist craftsmen from Central Asia and later from Nepal and northwest India, all of which were then Buddhist lands. Some cast images from this first Buddhist period may survive in Lhasa. After 842, central Tibet dissolved into political chaos for over 100 years, and from the 10th century onward the cultural initiative passed to a line of kings in western Tibet. For temple decorations, such as wood carving of doorways and posts, decorative painting on ceilings and woodwork, temple frescoes, and terra-cotta and stucco images, they drew heavily on the cultural resources of pre-Islāmic Kashmir. Surviving monasteries and temples, with their magnificent contents, were made known to the Western world in the 1930s. With the establishment of religious hegemonies in central Tibet from the 11th century onward, cultural contacts with Nepal and the Buddhist centres in the main Ganges Valley flourished as never before. Conversely, cultural contacts with China dwindled for several centuries, at least in central and southern Tibet. From this time until the 20th century, Tibetan religious art and Nepalese Buddhist art remained a single unified tradition. Meanwhile, eastern Tibet, where the ancient pre-Buddhist crafts of metalwork had never died out, began to develop religious styles under the influence of craftsmen from central Tibet. From that time, the spread of Tibetan culture and art became coterminous with the spread of Tibetan religion; and, thus, from the 13th century onward, when Tibetan lamas began to convert the Mongols, Mongolian religious art developed as a branch of Tibetan art. Through the Mongols, China began to extend its political influence over Tibet, and this led to a steady increase in Chinese cultural influence, especially in the east. From 1721, when the Chinese emperors became the suzerains of Tibet, Chinese influence was felt much more strongly throughout central as well as eastern Tibet, and Tibetan religious paintings and especially domestic decoration reveal distinct Chinese features.

Chinese
cultural
influence

Bronze statuette of Yamāntaka, the ferocious form of the Tibetan deity of wisdom, Mañjuśrī, who waged war against Yama, king of hell; Tibetan. In the American Museum of Natural History, New York.

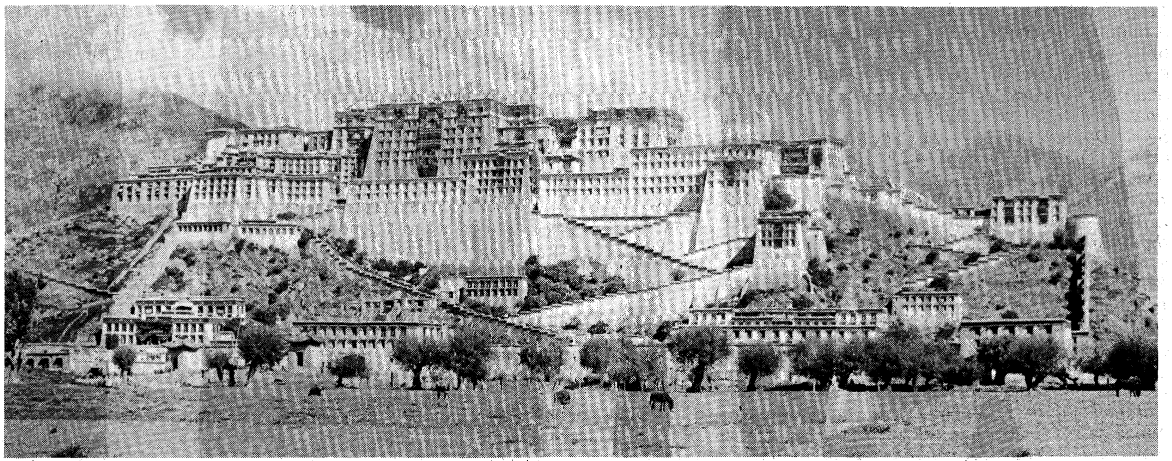
By courtesy of the American Museum of Natural History, New York

Decorative arts. In the main temple (*fo-khang*) of Lhasa there is a pre-Buddhist silver jug with a long neck surmounted by a horse's head; and there are textual references to all kinds of articles made of gold: a large golden goose holding seven gallons of wine, a wine vase, a miniature city decorated with gold lions, and golden bowls. Gold animals are mentioned as decorating the camp of King Ral-pa-can when a Chinese envoy visited him in 821. These early Tibetan skills lived on through the Buddhist period. Tibetan metalworkers have excelled in producing fine things for ritual and domestic use: ritual lamps, vases, bowls, bells, prayer wheels, decorated trumpets and horns, for the temples, and, for home use, ornamented teapots, jars, bowls, ladles, and especially beautiful stands, often in silver or gold, to hold porcelain teacups, capped by finely worked lids of precious metals. Hand-woven rugs of magnificent Central Asian and Chinese designs, always adapted to Tibetan preferences, cover low seats, and tables and cabinets of carved and painted wood were commonplace in prosperous homes.

Sculpture. From the 7th to 9th centuries there survive pre-Buddhist carved-stone pillars decorated with Chinese, Central Asian, and Indian motifs and also a stone lion showing traces of Persian influence.

Metal
casting

The art of casting images in bronze and other metals entered Tibet from Nepal and India. Having first followed foreign models, the Tibetans gradually developed their own styles and began to depict their own lamas and teachers as well as the vast pantheon of buddhas, gods, and goddesses inherited from India, each distinguished iconographically by posture, hand gestures, and accoutrements.



The Potala Palace, Lhasa, Tibet, 17th century.
Paul Popper (Popperfoto) Ltd.

(Of lesser divinities and especially of lamas, the identification is often difficult. It is rare that an image is named in an inscription and even rarer to find a date. Because of the extremely conservative nature of Tibetan art, correct dating within several centuries is often impossible.) Images of vast size, rising up through two or three stories, are quite often seen in Tibetan temples, and their construction and dedication is considered a work of vast religious merit.

Since images are mainly cast or molded, carving is restricted to decorative motifs, especially on wooden pillars and roof beams. Wood carving and terra-cotta, particularly in western and southern Tibet, were common. Papier-mâché, elaborately painted, was also used for masks of divinities. This use presumably originated in Kashmir.

By courtesy of O. Burchard, Berlin

Tibetan painting of Lcang-skya Rol-pa'i rdo-rje. The saint carries a book, one of the attributes of his patron, Tsong-kha-pa, founder of the Yellow Hat sect of Tibetan Buddhism. Chinese influence is shown by the architecture of the monasteries. In a private collection.

Painting: frescoes and temple banners. Temple interiors are usually covered with frescoes and often hung with painted banners, or tanka (*thang-ka*). For the preparation of the latter, a taut cotton cloth is impregnated with a mixture of chalk and glue, rubbed smooth by some suitable object; for example, a flat polished stone. A religious painter trained in the tradition draws in the outline, often using printed designs for the main figures. There is no scope for originality so far as the iconographic details of divinities are concerned, and, thus, such painting is a highly skilled craft. For decorative details—for example, flowers, cloud effects, rocks, and groups of devotees—there is wider scope. The tradition of fresco painting and temple banners certainly goes back to that of the great Buddhist monasteries of northwest India and the Ganges Valley, but these Indian origins of the 9th to 12th centuries are now entirely lost. The Indian Buddhist paintings of Ajanta are of a much earlier period (up to the 6th century AD), thus predating the great increase in the Buddhist pantheon and in occult symbolism typical of the later Indian Buddhism received by the Tibetans. Central Asian styles certainly reached central Tibet well before the 9th century, but, after that date, it was India and Nepal that were to have lasting influences on the development of Tibetan art. In more recent times, especially from the 18th century onward, Chinese influence became noticeable in the details of paintings, particularly in the freer but still balanced arrangement of the main figures and the use of Chinese-style landscapes as subsidiary decoration. With the disappearance of Buddhism from Central Asia and India from the 12th century onward, Tibetan art developed as a style exclusive to the Tibetans, the Newari Buddhists of the Nepal Valley, and the Tibetan converts of Mongolia.

Influence of Indian Buddhist painting

Decorative architectural motifs. For temples, monasteries, and official residences such as the Potala Palace of the Dalai Lama in Lhasa, the Tibetans used their own solid indigenous styles but embellished these with Indian, Nepalese, and (very much later) Chinese motifs. Tiered, ornamented temple roofs are of Indian origin, as received through Nepal and later through China. The magnificent interior carving is of Indian and Nepalese inspiration.

(D.L.S.)

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(T.T.R./M.S.S./T.V.W./S.Kh./D.L.S./J.R.Kr./F.I.)

Cervantes

Cervantes is the most celebrated of all Spanish writers. His novel *Don Quixote* has been translated, in whole or in part, into more than 60 languages. Editions continue regularly to be printed, and critical discussion of the work has proceeded unabated since the 18th century. At the same time, owing to their widespread representation in art, drama, and film, the figures of Don Quixote and Sancho Panza are probably familiar visually to more people than any other fictitious characters in world literature. Cervantes was a great experimenter. He tried his hand in all the major literary genres save the epic. He was a notable short-story writer, and a few of those in his collection of *Novelas exemplares* ("Exemplary Tales") attain a level close to that of *Don Quixote,* on a miniature scale.

Origins and early life

Miguel de Cervantes Saavedra was born in Alcalá de Henares, some 20 miles from Madrid, probably on Sept. 29 (the day of San Miguel), 1547. He was certainly baptized on October 9. He was the fourth of seven children in a family whose origins were of the minor gentry but which had come down in the world. Miguel's father was a barber-surgeon who set bones, performed bloodlettings, and attended lesser medical needs. The family moved from town to town, and little is known of Cervantes' early education. The supposition, based on a passage in one of the *Novelas exemplares,* that he studied for a time under the Jesuits, though not unlikely, remains conjectural. Unlike most Spanish writers of his time, including some of humble origin, he apparently did not go to a university. What is certain is that at some stage he became an avid reader of books. The head of a municipal school in Madrid, a man with Erasmist intellectual leanings named Juan López de Hoyos, refers to a Miguel de Cervantes as his "beloved pupil." This was in 1569, when the future author was 21, so—if this was the same Cervantes—he must either have been a pupil-teacher at the school or have studied earlier under López de Hoyos. His first published poem, on the death of Philip II's young queen, Elizabeth of Valois, appeared at this time.

Soldier and slave. That same year he left Spain for Italy. Whether or not this was because he was the "student" of the same name wanted by the law for involvement in a wounding incident is another mystery; the evidence is contradictory. In any event, in going to Italy Cervantes was doing what many young Spaniards of the time did to further their careers in one way or another. It seems that for a time he served as chamberlain in the household of Cardinal Giulio Acquaviva in Rome. However, by 1570 he had enlisted as a soldier in a Spanish infantry regiment stationed in Naples, a possession of the Spanish crown. He was there for about a year before he saw active service.

Confrontation at sea

Relations with the Ottoman Empire under Selim II were reaching a crisis, and the Turks occupied Cyprus in 1570. A confrontation between the Turkish fleet and the naval forces of Venice, the papacy, and Spain was inevitable. In mid-September 1571 Cervantes sailed on board the *Marquesa,* part of the large fleet under the command of Don Juan de Austria that engaged the enemy on October 7 in the Gulf of Lepanto near Corinth. The fierce battle ended in a crushing defeat for the Turks that was ultimately to break their control of the Mediterranean. There are independent accounts of Cervantes' conduct in the action, and they concur in testifying to his personal courage. Though stricken with a fever, he refused to stay below and joined the thick of the fighting. He received two gunshot wounds in the chest, and a third rendered his left hand useless for the rest of his life. He always looked back on his conduct in the battle with pride. From 1572 to 1575, based mainly in Naples, he continued his soldier's life; he was at Navarino and saw action in Tunis and La Goleta. He must also, when opportunity offered, have been familiarizing himself with Italian literature. Perhaps with a recommendation for promotion to the rank of captain, more likely just leaving the army, he set sail for Spain in September 1575 with letters of commendation to the King from the Duque de Sessa and Don Juan himself.

Treatment in captivity

On this voyage his ship was attacked and captured by Barbary corsairs, and Cervantes, together with his brother Rodrigo, was sold into slavery in Algiers, the centre of the Christian slave traffic in the Muslim world. The letters he carried magnified his importance in the eyes of his captors. This had the effect of raising his ransom price, and thus prolonging his captivity, while also, it appears, protecting his person from punishment by death, mutilation, or torture when his four daring bids to escape were frustrated. His masters, the renegade Dali Mami and later Hasan Paşa, treated him with considerable leniency in the circumstances, whatever the reason. At least two contemporary records of the life led by Christian captives in Algiers at this time mention Cervantes. He clearly made a name for himself for courage and leadership among the captive community. At long last, in September 1580, three years after Rodrigo had earned his freedom, Miguel's family, with the aid and intervention of the Trinitarian friars, raised the 500 gold escudos demanded for his release. It was only just in time before Hasan Paşa sailed for Constantinople, taking his unsold slaves with him. Not surprisingly, this, the most adventurous period of Cervantes' life, supplied subject matter for several of his literary works, notably the Captive's tale in *Don Quixote* and the two Algiers plays, *El trato de Argel* ("The Traffic of Algiers") and *Los baños de Argel* ("The Bagnios of Algiers"), as well as episodes in a number of other writings, although never in straight autobiographical form.

Economic hardships

Civil servant and writer. Back in Spain, most of the rest of his life was to be spent in complete contrast with his decade of action and danger. He would be constantly short of money and in tedious and exacting employment; it would be 25 years before he scored a major literary success with *Don Quixote.* On his return home he found that prices had risen and the standard of living for many, particularly those of the middle class like his family, had fallen. The euphoria of Lepanto was a thing of the past. Cervantes' war record did not now bring the recompense he expected. He applied unsuccessfully for several administrative posts in Spain's American empire. The most he succeeded in acquiring was a brief appointment as royal messenger to Oran in 1581. He followed Philip II and the court to Lisbon in newly annexed Portugal in vain.

First published fiction

Around this time he had an affair with a young married woman called Ana de Villafranca (or Ana Franca de Rojas), the fruit of which was a daughter. Isabel de Saavedra, Cervantes' only child, was later brought up in her father's household. Late in 1584 he married Catalina de Salazar y Palacios, 18 years his junior. She had a small property in the village of Esquivias in La Mancha. Little is known about their emotional relationship. There is no reason to suppose that the marriage did not settle down into an adequate companionableness, despite Miguel's enforced long absences from home. Neither is there any special reason to suppose that Catalina was an inspiration or a model for characters in the poetry Cervantes was now writing or in his first published fiction, *La Galatea* (1585), in the newly fashionable genre of the pastoral romance. The publisher, Blas de Robles, paid him 1,336 reales for it, a good price for a first book. The dedication of the work to Ascanio Colonna, a friend of Acquaviva, was a bid for patronage that does not seem to have been productive. Doubtless helped by a small circle of literary friends, such as the poet Luis Gálvez de Montalvo, the book did bring Cervantes' name before a sophisticated reading public. But the only later editions in Spanish to appear in the author's lifetime were those of Lisbon, 1590, and Paris, 1611. *La Galatea* breaks off in mid-narrative; judging by his repeatedly expressed hopes of writing a sequel, Cervantes evidently maintained a lasting fondness for the work.

Failure as a dramatist

He also turned his hand to the writing of drama at this time, the early dawn of the Golden Age of the Spanish theatre. He contracted to write two plays for a theatrical manager called Gaspar de Porras in 1585, one of which, *La confusa* ("Comedy of Confusion"), he later described as the best he ever wrote. Many years afterward he claimed to have written 20 or 30 plays in this period, which, he noted, were received by the public without being pelted with vegetables or booed off the stage. The number is vague; only two certainly survive from this time, the historical tragedy of *La Numancia* ("Numantia") and *El trato de Argel.* He names nine plays, the titles of a few of which sound like the originals of plays reworked and published years later in the collection *Ocho comedias, y ocho entremeses nuevos* (1615; "Eight New Plays and Eight New Interludes"). Fixed theatres were just becoming established in the major cities of Spain, and there was an expanding market geared to satisfying the demands of a public ever more hungry for entertainment. Lope de Vega was about to respond to the call, stamping his personal imprint on the Spanish *comedia* and rendering all earlier drama, including that of Cervantes, old-fashioned or inadequate by comparison. Though destined to be a disappointed dramatist, Cervantes went on trying to get managers to accept his stage works. By 1587 it was clear that he was not going to make a living from literature, and he was obliged to turn in a very different direction.

Cervantes became a commissary of provisions for the great Armada. Requisitioning corn and oil from grudging rural communities was a thankless task, but it was at least a steady job, with a certain status. It took him traveling all over Andalusia, an experience he was to put to good use in his writing. He was responsible for finances of labyrinthine complexity, and the failure to balance his books landed him in prolonged and repeated trouble with his superiors. There also was constant argument with municipal and church authorities, the latter of which more than once excommunicated him. The surviving documentation of the accountancy and negotiations involved is considerable.

After the disastrous defeat of the Armada in 1588 he gravitated to Seville, the commercial capital of Spain and one of the largest cities in Europe. In 1590 he applied to the Council of the Indies for any one of four major crown posts vacant in Central and South America. His petition was curtly rejected. Wrangles over his accounts and arrears of salary dragged on. He seems to have kept some contact with the literary world; there is a record of his buying certain books, and he must have managed to find time for reading. In 1592 he signed a contract to supply six plays to a theatrical manager, one Rodrigo Osorio. Nothing came of this. His commissary work continued, and the litigation came to a head; in September 1592 he was imprisoned for a few days in Castro del Río.

In 1594 he was in Madrid seeking a new post. He received an appointment that took him back to Andalusia to collect overdue taxes. Although it was in effect a promotion, the job was no more rewarding than the previous one and was similarly fraught with financial difficulties and confrontations. Cervantes was not by temperament a businessman. Probably by mutual agreement the appointment was terminated in 1596. The previous year he had won first prize (three silver spoons) in a poetry competition in Zaragoza. Back in Seville, it is likely that at about this time he started seriously writing stories, not to mention a wickedly satirical sonnet on the conduct of the Duque de Medina Sidonia, to be followed by one obliquely disrespectful of the recently deceased king himself. Again he met with financial troubles. In the summer of 1597 discrepancies in his accounts of three years previous landed him in the Crown Jail of Seville. He was confined until the end of April 1598 and perhaps conceived there the idea of *Don Quixote,* as a remark in the first prologue suggests.

Information about Cervantes' life over the next four or five years is sparse. He had left Seville, and, perhaps for a while in Esquivias and Madrid, later for certain in Valladolid (where the royal court established itself from 1601 to 1606), he must have been writing the first part of *Don Quixote.* Early versions of two of his stories, *Rinconete y Cortadillo* ("Rinconete and Cortadillo") and *El celoso extremeño* ("The Jealous Estremaduran"), found their way into a miscellaneous compilation, unpublished, made by one Francisco Porras de la Cámara.

Publication of *Don Quixote*

In July or August 1604 Cervantes sold the rights of *El ingenioso hidalgo Don Quixote de la Mancha* ("The Ingenious Hidalgo Don Quixote of La Mancha"; known as *Don Quixote,* Part I) to the publisher-bookseller Francisco de Robles for an unknown sum. License to publish was granted in September and the book came out in January 1605. There is some evidence of its content's being known or known about before publication—among others, to Lope de Vega, the vicissitudes of whose relations with Cervantes were then at a low point. The compositors at Juan de la Cuesta's press in Madrid are now known to have been responsible for a great many errors in the text, many of which were long attributed to the author. The novel was an immediate success, though not as sensationally so as Mateo Alemán's *Guzmán de Alfarache,* Part I, of 1599. By August 1605 there were two Madrid editions, two published in Lisbon, and one in Valencia. There followed those of Brussels, 1607; Madrid, 1608; Milan, 1610; and Brussels, 1611. Part II, *Segunda parte del ingenioso cavallero Don Quixote de la Mancha* ("Second Part of the Ingenious Knight Don Quixote of La Mancha"), came out in 1615. Thomas Shelton's English translation of the first part appeared in 1612. The name of Cervantes was soon to be as well known in England, France, and Italy as in Spain.

The sale of the publishing rights, however, meant that he made no more financial profit on Part I of his novel. He had to do the best he could with patronage. The dedication to the young duque de Béjar had been a mistake. He had better fortune with two much more influential persons: the Conde de Lemos, to whom he would dedicate Part II and no less than three other works, and Don Bernardo de Sandoval y Rojas, archbishop of Toledo. This eased his financial circumstances somewhat. However, it is apparent that he would have liked a securer place in the pantheon of the nation's writers than he ever achieved during his lifetime—he wanted a reputation comparable to that enjoyed by Lope de Vega or the poet Luis de Góngora y Argote, for example. His sense of his own marginal position may be deduced from his *Viage del Parnaso* (1614; "Voyage to Parnassus"), two or three of the later prefaces, and a few external sources. Nevertheless, relative success, still unsatisfied ambition, and a tireless urge to experiment with the forms of fiction ensured that, at the age of 57, with less than a dozen years left to him, Cervantes was just entering the most productive period of his career.

No graciousness descended on his domestic life. A stabbing incident in the street outside the house in Valladolid, in June 1605, led ridiculously to the whole household's arrest. When they later followed the court to Madrid, he continued to be plagued by litigation over money and now, too, over Isabel's unedifying marital affairs. The family lodged in various streets over the next few years before finally settling in the Calle de León. Like a number of other writers of the day, Cervantes nursed hopes of a secretarial appointment with the Conde de Lemos when, in 1610, the Conde was made viceroy of Naples. Once more he was disappointed. He had joined a fashionable religious order, the Slaves of the Most Blessed Sacrament, in 1609, and four years later he became a Franciscan tertiary, which was a more serious commitment. We know, too, of some increased involvement in the literary life of the capital in the form of his attendance at the Academia Selvaje, a kind of writers' salon, in 1612.

The *Novelas exemplares*

The next year, the 12 *Novelas exemplares* were published. The prologue contains the only known authentic portrait of the author:

> of aquiline countenance, with dark brown hair, smooth clear brow, merry eyes and hooked but well-proportioned nose; his beard is silver though it was gold not 20 years ago; large moustache, small mouth with teeth neither big nor little, since he has only six of them and they are in bad condition and

> worse positioned, for they do not correspond to each other; the body between two extremes, neither tall nor short; a bright complexion, more pale than dark, somewhat heavy in the shoulder and not very light of foot.

Cervantes' claim in this prologue to be the first to write original novellas (short stories in the Italian manner) in Castilian is substantially justified. Their precise dates of composition are in most cases uncertain. There is some variety in the collection, within the two general categories of romance-based stories and realistic ones. *El coloquio de los perros* ("Colloquy of the Dogs"), a quasi-picaresque novella, with its frame tale *El casamiento engañoso* ("The Deceitful Marriage"), is probably Cervantes' most profound and original creation next to *Don Quixote.* In the 17th century the romantic stories were the more popular; James Mabbe chose precisely these for the selective English version of 1640. Nineteenth- and 20th-century taste has preferred the realistic ones, but the others are receiving again something like their critical due.

In 1614 Cervantes published *Viage del Parnaso,* a long allegorical poem in mock-mythological and satirical vein, with a postscript in prose. It was devoted to celebrating a host of contemporary poets and satirizing a few others. The author there admitted that writing poetry did not come easily to him. But he held poetry in the highest esteem as a pure art that should never be debased. Having lost all hope of seeing any more of his plays staged, in 1615 he had eight of them published, together with eight short comic interludes, in *Ocho comedias, y ocho entremeses nuevos.* The plays show no shortage of inventiveness and originality but lack real control of the medium. The entremeses, however, are reckoned among the very best of their kind.

Don Quixote, Part II

It is not certain when Cervantes began writing Part II of *Don Quixote,* but he had probably not gotten much more than halfway through by late July 1614. Around September a spurious Part II was published in Tarragona by someone calling himself Alonso Fernández de Avellaneda, an unidentified Aragonese, who was an admirer of Lope de Vega. The book is not without merit, if crude in comparison with its model. In its prologue the author gratuitously insulted Cervantes, who not surprisingly took offense and responded, though with relative restraint if compared with the vituperation of some literary rivalries of the age. He also worked some criticism of Fernández de Avellaneda and his "pseudo" Quixote and Sancho into his own fiction from chapter 59 onward.

Don Quixote, Part II, emerged from the same press as its predecessor late in 1615. It was quickly reprinted in Brussels and Valencia, 1616, and Lisbon, 1617. Parts I and II first appeared in one edition in Barcelona, 1617. There was a French translation of Part II by 1618 and an English one by 1620. The second book capitalizes on the potential of the first, developing and diversifying without sacrificing familiarity. Most people agree that it is richer and more profound.

Last works

In his last years Cervantes mentioned several works that apparently did not get as far as the printing press, if indeed he ever actually started writing them. There was *Bernardo* (the name of a legendary Spanish epic hero), the *Semanas del jardín* ("Weeks in the Garden"; a collection of tales, perhaps like the *Decameron*), and the continuation to his *Galatea.* The one that was published, posthumously in 1617, was his last romance, *Los trabaios de Persiles y Sigismunda, historia setentrional* ("The Labours of Persiles and Sigismunda: A Northern Story"). In it Cervantes sought to renovate the heroic romance of adventure and love in the manner of the *Aethiopica* of Heliodorus (*c.* AD 230–240). It was an intellectually prestigious genre destined to be very successful in 17th-century France. Intended both to edify and to entertain, the *Persiles* is an ambitious work that exploits the mythic and symbolic potential of romance. It was very successful when it appeared; there were eight Spanish editions in two years and French and English translations in 1618 and 1619, respectively.

In the dedication, written three days before he died, Cervantes, "with a foot already in the stirrup," movingly bade farewell to the world. Clear-headed to the end, he seems to have achieved a final serenity of spirit. He died almost certainly on April 22, 1616, not on the 23rd as had been traditionally thought. The burial certificate indicates that the latter was the day he was buried, in the convent of the Discalced Trinitarians in the Calle de Cantarranas (now the Calle de Lope de Vega). The exact spot is not marked. No will is known to have survived.

Importance and influence. Cervantes' greatest creation, *Don Quixote,* appeared at the end of more than a century of remarkably innovative and diverse achievement by Spanish prose fiction writers. It was immediately preceded by the sudden efflorescence of works in the new genre of the picaresque, in which Cervantes himself participated with *Rinconete y Cortadillo. Don Quixote,* however, parodied the romances of chivalry, which in published titles had outnumbered all other kinds of fiction in the 16th century, but which were now a genre in marked decline in Spain. The parody takes an unusual form. Its protagonist, an elderly gentleman addicted to reading these romances, goes out of his mind, believes them to have been historically true, and decides to become a knight-errant, to go out into the world and live his own romance of chivalry. The story of his career is presented in the form of a basically realistic novel. The formula proved to have a remarkable potential for exploitation.

Critical interpretation

The figures of Don Quixote, Sancho Panza, and Don Quixote's horse, Rosinante, immediately gripped the popular imagination, as their recorded appearances in festive parades from Peru to Germany between 1605 and 1617 testify. Contemporaries evidently did not take the book as seriously as later generations have done, but by the end of the century its stock had greatly risen, especially abroad. It came to be seen as a mock epic in prose, and the "grave and serious air" of the author's irony was much admired. The occasional esoteric interpretation of the work as disguised political or other satire began to appear. An increasingly sympathetic view of the hidalgo signaled the start of an interpretative transformation effected by the German Romantics. The ludicrous hero of the most comical of books became the tragic hero of the saddest. However, despite some distortion, Cervantes' novel now revealed its unsuspected depths. The way was opened to modern criticism, which has been remarkable for its diversity.

In the history of the modern novel the role of *Don Quixote* is recognized as seminal. Evidence of this may be seen in Defoe, Fielding, Smollett, and Sterne. Connections have been noticed with major works by a great many of the classic novelists of the 19th century, including Scott, Dickens, Flaubert, Pérez Galdós, Melville, and Dostoyevsky. And the same is true, in different ways, in the case of many post-realist writers of the 20th century, from Joyce to Borges. The explanation would seem to have less to do with "influence" than with Cervantes' having succeeded in tapping the core of the possibilities of the novel. In addition, *Don Quixote* has proved to be a remarkable source of inspiration to creators in other genres and media. From the 17th century onward, stage adaptations, operatic versions, ballets, and musical compositions have been made in considerable number, together with cinematic, television, and cartoon versions in the 20th century. *Don Quixote* has also inspired artists such as Hogarth, Goya, Daumier, and Picasso. The illustrations by Gustave Doré are perhaps the most celebrated.

MAJOR WORKS

PROSE FICTION: *Primera parte de la Galatea* (1585; *Galatea: A Pastoral Romance,* trans. by Gordon Willoughby James Gyll, 1867); *El ingenioso hidalgo Don Qvixote de la Mancha,* and *Segunda parte del ingenioso cavallero Don Qvixote de la Mancha* (1605, 1615; *The History of the Valorous and Wittie Knight-Errant Don Quixote of the Mancha,* trans. by Thomas Shelton, 1612–20); *Novelas exemplares* (1613; *Exemplarie Novells,* trans. by Don Diego Puede-Ser [James Mabbe], 1640); *Los trabaios de Persiles y Sigismunda, historia setentrional* (1617; *The Travels of Persiles and Sigismunda: A Northern History,* 1619).

DRAMA: *La Numancia* (1784; *Numantia, A Tragedy,* trans. by Gordon Willoughby James Gyll, 1870); *Ocho comedias, y ocho entremeses nuevos* (1615; *The Interludes of Cervantes,* trans. by S. Griswold Morley, 1948).

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(E.C.R.)

Charlemagne

Charlemagne (Carolus Magnus, Charles the Great) as king of the Franks (768–814) conquered the Lombard kingdom in Italy, subdued the Saxons, annexed Bavaria to his kingdom, fought campaigns in Spain and Hungary, and, with the exception of the Kingdom of Asturias in Spain, southern Italy, and the British Isles, united in one superstate practically all the Christian lands of western Europe. In 800 he assumed the title of emperor. (He is reckoned as Charles I of the Holy Roman Empire, as well as Charles I of France.) Besides expanding its political power, he also brought about a cultural renaissance in his empire. Although this *imperium* survived its founder by only one generation, the medieval kingdoms of France and Germany derived all their constitutional traditions from Charles's monarchy. Throughout medieval Europe, the person of Charles was considered the prototype of a Christian king and emperor.

Foundation of the dynasty

Early years. Charles was born probably in 742 (on April 2), the elder son of Pepin III, also called Pepin the Short. Pepin and his older brother, Carloman, had just jointly assumed the government of the Frankish kingdom as *maior domus,* or "mayor of the palace." The dynasty, later called Carolingian after Charlemagne, had originated in the Meuse-Moselle region on the borders of modern France, Germany, Belgium, and The Netherlands. In the course of a few generations, it had, as mayors of the palace to the Merovingians, gained control of the entire Frankish kingdom. Charlemagne's grandfather, Charles Martel, reconstituted a realm that had been on the point of breaking up, and, without infringing on the royal prerogatives of the otherwise powerless Merovingians, he had in effect bequeathed the empire to his sons, Pepin and Carloman, like a family inheritance.

Alliance with the papacy

Charles grew to manhood while his father was engaged in acquiring sole sovereignty and the kingship. On Carloman's retirement to a monastery, Pepin eliminated the latter's sons from the government. Having thus prepared the way, he had himself proclaimed king in 751, after dethroning the Merovingians. An oracular response by Pope Zacharias furnished the ecclesiastical approbation for thus shunting aside the former reigning house, which had been held sacred. Zacharias' successor, Stephen II, arrived in the Frankish kingdom during the winter of 753–754, in order to seek help against the Lombards who were attacking Rome. As the reigning monarch's oldest son, Charles, then about 12 years of age, travelled ahead to welcome the Pope, who anointed him king, along with his father and his brother Carloman, thus sanctioning the new dynasty. The political alliance between the Franks and the Pope against the Lombards was affirmed on the same occasion. When his father subdued Aquitaine (France south of the Loire) in a series of yearly campaigns beginning in 760, reasserting the integrity of the Frankish kingdom all the way to the Pyrenees, Charles repeatedly accompanied the army.

These youthful experiences probably contributed to the formation of Charles's character and to the formulation of his aims. He shared with his father an unbending will to power, a readiness to fight resolutely against external enemies and to increase his domains, and the determination to rule by himself even if it meant usurping the rights of close relatives. Charles early acknowledged the close connection between temporal power and the church; he had a high regard for the church and the king's duty to spread the Christian faith and, while asserting royal suzerainty over the church, considered himself accountable to God for the Christians entrusted to him.

King of the Franks. In accordance with old Frankish custom, the kingdom was divided on Pepin's death in 768 between his two sons. It was not long, however, before a strong rivalry sprang up between the brothers: with his mother's support, Charles concluded, with the Lombard king Desiderius, whose daughter he married, and with his cousin Duke Tassilo of Bavaria, alliances directed against Carloman.

Expansion in Italy

On Carloman's sudden death in 771, Charles was able to make himself sole ruler of the kingdom, unopposed by his young nephews, whose rights he ignored. When Carloman's widow with her children and a few remaining supporters had fled to the Lombard court, and King Desiderius, breaking his alliance with Charles, put pressure on the Pope to anoint Carloman's sons as Frankish kings, Charles was forced to come to the aid of Pope Adrian I. He marched on the Lombard capital, Pavia, and after its fall made himself king of the Lombards. His brother's sons, who had fallen into his hands, disappeared. While the siege of Pavia was still in progress, Charles journeyed to Rome, where he celebrated Easter 774 with the Pope and reiterated, in St. Peter's Basilica, his father's promise to transfer to papal rule large sections of Italy. But he actually enlarged the Pope's lands only slightly, assuming for himself the sovereignty over the entire Lombard kingdom.

By courtesy of Weidenfeld & Nicholson Ltd.; photograph, Ann Munchow

Charlemagne, stucco statue, probably 9th century. Church of St. John the Baptist, Münster, Switz.

Charles had fought the pagan Saxons, in what is now

Lower Saxony and Westphalia, in retribution for their attacks on the lower Rhine region, as early as 772, before the first Italian campaign. From 775 on, however, it was his goal to subdue the whole Saxon tribe, converting it to Christianity and integrating it into his kingdom. This aim appeared to have been realized after several campaigns culminating in declarations of allegiance by the Saxon nobility and mass baptisms performed in 775–777. A diet held in 777 in Paderborn sealed the submission of the Saxons. Among those attending the diet had been some Arab emissaries from northern Spain who sought Charles's aid in their uprising against the Umayyad *amir* of Córdoba. In the summer of 778 Charles advanced into Spain and laid siege to Saragossa, without, however, being able to take the city. Retreating across the Pyrenees, the Frankish army was badly mauled by the Basques. Roland, warden of the Breton march, who died on this occasion, was later immortalized in legend and poetry.

This defeat marks the end of the first period of Charles's rule, the period of vigorous expansion. Within a decade he had become the sole ruler of the Franks, conquered the Lombard kingdom, visited Rome, subdued the Saxons, invaded Spain. Henceforth he was concerned with defending and safeguarding his quickly won gains (which were to be extended only on the right bank of the Rhine), while consolidating the state internally and protecting cultural life and the rule of law.

Subjugation of the Saxons

Not long after Charles's defeat in Spain, the Saxons rose up once more. The war against them became the longest and most cruel war fought by the Franks. In Charles's eyes, the resistance of this people that had undergone baptism and signed a treaty of allegiance amounted to political high treason and religious apostasy. These offenses called for severe punishment, and 4,500 Saxons were reported to have been executed en masse in 782. New outbreaks occurred after 792, and the last Saxons were not vanquished until 804. Between 772 and 804, Charles took the field against the Saxons no fewer than 18 times. In the end he carried out his aim of not only subjecting them to his rule but also incorporating them fully into his empire. Given the indissoluble tie between temporal power and the Christian faith, this meant they had to be converted. But the violent methods by which this missionary task was carried out had been unknown to the earlier Middle Ages, and the sanguinary punishment meted out to those who broke canon law or continued to engage in pagan practices called forth criticism in Charles's own circle, for example by Alcuin, his adviser and head of his palace school.

When, in 788, Charles deposed his cousin Duke Tassilo III of Bavaria, who had acknowledged the Frankish kings as feudal lords, he in effect deprived of its independence the last of the German tribes beyond the Rhine. The Bavarians, who had long been Christians, were now directly integrated into the empire. The West Germanic tribes of the Alemanni, Bavarians, Saxons, and Thuringians thus found themselves for the first time gathered into one political unit. Charles's conquests on the right bank of the Rhine were, however, not limited to the Germanic tribes. Making Ratisbon (Regensburg), the residence of the Bavarian dukes, his base, he conducted several campaigns, partly under his own command, against the Avar kingdom (in modern Hungary and Upper Austria). The remaining Avar principalities and the newly founded Slav states of the Danubian region drifted into a loose dependence on the Franks, whose sovereignty they more or less acknowledged.

The gigantic expansion of the Frankish state, raising it far above the tribal states of the early Middle Ages, entailed qualitative as well as quantitative changes. Yet the idea of bestowing on Charles the Roman title of emperor arose only at a very late stage and out of a specific political constellation. While the Eastern, or Byzantine, Empire laid claim to universal recognition, the popes, constitutionally still subjects of Byzantium, were opposed to the iconoclastic religious policies of the Eastern emperors. Moreover, under the protection of Charles, Pope Adrian sought to erect an autonomous domain over central Italy, the more so as the Byzantines, abandoning for all practical purposes Rome and Ravenna, were asserting their rule only in Sicily and the southernmost edge of Italy. The papacy's desire for independence found a significant expression in the Donation of Constantine, a forgery dating probably from the first few years of Adrian's reign and purporting to legitimize these papal aims in the name of the first Christian emperor, Constantine I the Great. Charles paid a second visit to Rome in 781, when he had the Pope crown his young sons Pepin and Louis as kings of the Lombards and Aquitanians and gained de facto recognition of his Italian position from the Byzantine empress Irene, the mother of Constantine VI. The entente that existed between Charles and Byzantium came to an end after a Frankish attack on southern Italy in 787.

Emperor of the West. In the end, local Roman conflicts brought about the clarification of the city's constitutional position. In May 799, Pope Leo III was waylaid in Rome by personal enemies. He took refuge at the court of Charles, who had him conducted back to the city and who in November 800 came to Rome himself, where he was received with imperial honours. Before Charles and a synod, Pope Leo cleared himself under oath of the charges made by his enemies. During Christmas mass in St. Peter's, the Romans acclaimed Charles emperor, whereupon the Pope crowned and perhaps anointed him.

Coronation in Rome

The imperial title was by nature a Roman dignity. While the acclamation represented the juridically conclusive act, it was the coronation at the hands of the Pope that, though of no constitutional importance, was to acquire for the Franks great significance. The Pope had been determined to make Charles emperor, deciding to a large extent the outward form; yet Charles was surely not surprised by these events. His famous statement quoted by one of his favourites, the Frankish historian Einhard, that he would not have set foot in church that Christmas if he had known the Pope's intention, implies a criticism of the ceremony initiated by the Pope, as well as a formal expression of humility. The crowning had been preceded by negotiations. While Charles's imperial rank was legally substantiated by the fact of his dominion over the western part of the old Roman Empire, the desire to counteract the petticoat rule of the empress Irene (who had dethroned and blinded her son in 797) also played a role. Residing in Rome four months and pronouncing sentence on the Pope's enemies as rebels guilty of lese majesty, Charles grasped the imperial reins with a firm hand. Likewise, after his return to Aachen (Aix-la-Chapelle), he promulgated laws in full consciousness of his rank as emperor.

Byzantium braced itself for the usurper's attack, but Charles merely wished to see his new rank and his dominion over Rome recognized in negotiations; he gained his point in 812 when the emperor Michael I acknowledged him as emperor, though not as emperor of the Romans. While the imperial title did not bring Charles any additional powers, his control of Rome was now legitimized, and the estrangement of the papacy from Byzantium and its rapprochement with the Franks, a major historical event that had been initiated in 754, was rendered incontrovertible. A significant result of this development was the tradition to which Charles's assumption of the imperial title and function gave rise: all medieval concepts of empire and all the bonds between the constitutional traditions of the Franks and the later Holy Roman Empire with the Roman Empire founded by Augustus were based on the precedent of Charles's imperial title and position.

Court and administration. The creation of the empire was chiefly legitimized by Charles's efforts to raise its cultural level internally. When Charles came to power, the Frankish kingdom's cultural, administrative, and legal institutions were still relatively undeveloped. The Frankish king, for example, possessed no permanent residence. In the summer months he travelled about, deciding political issues and dispensing justice in assemblies of spiritual and temporal lords; above all, summer was the season for military campaigns. During the winter, from Christmas to Easter and sometimes longer, the king lived and held court at one of the imperial palaces. Charles especially favoured those situated in the Frankish heartland: only rarely did he spend the winter in one of the newly won territories, in encampment in Saxony, in Ratisbon, or in Rome. Not

Charles's peripatetic court

until 794 did Aachen, which the aging monarch liked because of its warm springs, become the court's abode, indeed almost a residence, during every winter and often even in summer. Here Charles built, partially with materials imported from Rome and Ravenna, the court church that is still standing, as well as the palace whose walls were incorporated into the 14th-century city hall.

Charles's court consisted of his family, of the clergy in his personal service, who were called the king's *capella,* and of temporal officials, among them the count palatine, the seneschal, and the master of the royal household. These men were occasionally joined, on an informal basis, by other spiritual or temporal men of rank who spent some time in the ruler's presence. For Charles had the ambition to make his court the intellectual, as well as the political and administrative, centre of the realm and accordingly summoned prominent scholars from all parts of the empire and even from abroad. Among these the most important were Einhard and Alcuin.

With the help of these and other literary men, Charles established a court library containing the works of the Church Fathers and those of ancient authors, and he founded a court academy for the education of young Frankish knights. Last but not least, he himself took part with his family and the learned and lay members of his entourage in a cultivated social life that afforded him entertainment no less than instruction. His mother tongue was an Old High German idiom, besides which he presumably understood the Old French dialect spoken by many Franks; as a grown man, he also learned Latin and some Greek, had historical and theological writings, including St. Augustine's *City of God,* read aloud to him, and acquired a rudimentary knowledge of mathematics and astronomy.

The Carolingian renaissance

The court's cultural interests, however, extended beyond the intellectual gratification of a small circle, such as the exchange of verses and letters. Efforts were also made to raise the level of religious observance, morality, and the process of justice throughout the empire. The clearest and most famous instance of this was the *Epistula de litteris colendis,* dating presumably from 784 to 785 and compiled in Charles's name by Alcuin. Its main argument lies in the assertion that the right faith—indeed, every right thought—must be clothed in the appropriate form and language, lest it be falsified; hence, the prescription of intensive study of Latin language and literature for all monastic and cathedral schools. The spiritual and literary movement called the "Carolingian renaissance" had many centres, especially in the empire's monasteries; but it cannot be evaluated without reference to Charles's court and to his endeavour to call on the best minds of the whole world, setting them to work in the education of the clergy and, in the final instance, of the whole people. The court's theological knowledge and intellectual self-confidence are reflected in the *Libri Carolini,* a comprehensive treatise written about 791 in Charles's name and directed against the Council of Nicaea (787), at which Greeks and papal plenipotentiaries had countenanced the practice of iconolatry; at the same time, the *Libri Carolini* did not spare the iconoclasts.

Through this court, Charles ruled and administered his empire and dispensed justice. Once or twice a year at least, the court and the chief magistrates and nobles from all parts of the empire joined in a general assembly held either in the Frankish heartland or in one of the conquered territories. It is indicative of the unique structure of the Carolingian Empire that one cannot draw clear distinctions between an assembly of the armed forces, a constitutional assembly of the nobility, and a church synod: juridical, military, and ecclesiastical affairs were invariably discussed at one and the same time by the representatives of the nobility and the clergy. Above them all towered the figure of Charlemagne.

Internal administration

On the local level the ruler was represented in every region by counts and bishops. Liaison between these personages and the court was maintained through royal messengers who travelled about at Charles's command, usually in pairs made up of a civil servant and a clerical dignitary. Royal commands did not have to be written out, although Charles's decrees (capitularies) increasingly came to be recorded in writing, at first rather imprecisely, in the last two decades of his reign; the forms coined by the "renaissance" gained ground only with time. Charles respected the traditional rights of the various peoples and tribes under his dominion as a matter of principle, and, after he became emperor, he had many of them recorded. The capitularies served partly as complements to tribal laws, partly as regulations applying to the most disparate aspects of public and private life, and in part also as specific instructions issued to royal messengers, counts, bishops, and others. Punitive decrees against highwaymen, dispositions concerning military levies, orders for the people to take an oath of allegiance to the emperor or to teach all Christians to recite the Lord's Prayer, are found intermingled in the capitularies with jurisdictional dispositions and regulations about the internal organization of monasteries; temporal and spiritual problems are rarely treated separately. Taken as a whole, the legal documents of Charles's reign bear witness to a great concern, born of profound moral and religious convictions, with the administration of justice and with public enlightenment, but they also show discrepancies between the ideal and reality.

Conspiracies

Limitations of his rule. Charles's organization of the empire was, however, not without its defects and limitations. The sovereign's power was restricted only by theoretical principles of law and custom, not by institutions or countervailing forces. Significantly, the records report little about opposition movements and conspiracies, which, in fact, did exist. A rebellion that Thuringian counts launched against Charles in 786 can perhaps be explained as ethnic opposition to the centralism of the Franks. More ominous was an aristocratic conspiracy that in 792 attempted to place on the throne the hunchback Pepin, Charles's only son from his first marriage, which was later declared invalid; yet here, too, the political concepts and motives remain unknown. These events and, more clearly still, the history of the empire under Charles's successor, Louis, show the extent to which the political system had been designed for one person on whose outstanding abilities everything depended and with whose disappearance it threatened to collapse. Their self-confidence enhanced by Charles's educational policy, the clergy could not accept for all time his theocracy without opposing it with their own political and religious principles. The temporal nobility that had built the empire with the Carolingians could be firmly tied to the dynasty only as long as new conquests held out the prospect of new spoils and fiefs; if these failed to materialize, there remained only the care of one's properties in the different regions and the hope of gaining advantages from party strife. External expansion, however, could not advance substantially beyond the borders reached by 800; in fact, economic and technical resources were insufficient to hold together and administer what had already been won and to defend it against foreign enemies. Charles's empire lacked the means by which the Romans had preserved theirs: a money economy, a paid civil service, a standing army, a properly maintained network of roads and communications, a navy for coastal defense. Already in Charles's lifetime, the coasts were being threatened by the Normans. In 806 Charles planned a division of the empire between his sons, but after the death of the elder two he crowned Louis of Aquitaine his coemperor and sole successor at Aachen in 813. It was only a few months later that Charles himself died there on January 28, 814.

Personality and influence. Charlemagne's posthumous fame shone the more brightly as the following generations were unable to preserve the empire's internal peace, its unity, and its international position. Even after the Carolingian dynasty had become extinct, political tradition in the East Frankish (German) kingdom and empire, as well as in the West Frankish (French) kingdom, drew sustenance from the example set by Charlemagne. Under Otto I, Aachen became the city in which the rulers of Germany were crowned, and, at Frederick I Barbarossa's request, the antipope Paschal III canonized Charlemagne in 1165. In France the Capetians, beginning with Philip II Augustus, revived the traditions that had grown up around Charlemagne. The controversial question whether the Germans or the French were the true successors of Charlemagne was kept alive through the Middle Ages and into modern times. Napoleon called himself Charlemagne's successor; after the end of World War II, discussions of a united, Christian, "occidental" Europe invoked his model. Hand in hand with these political traditions went those in popular legend and poetry, culminating in the Roland epics. Nor did Charlemagne's fame stop at the boundaries of what was once his empire; some Slavic languages derived their term for "king" from his name (Czech *král,* Polish *król,* etc.).

Charles left no biographical document; his personality can be constructed only from his deeds and the reports left by contemporaries. This is how Einhard, who lived at the court from about 795 on, described Charlemagne's character and appearance in his famous *Vita Karoli Magni:* "He had a broad and strong body of unusual height, but well-proportioned; for his height measured seven times his feet. His skull was round, the eyes were lively and rather large, the nose of more than average length, the hair gray but full, the face friendly and cheerful. Seated or standing, he thus made a dignified and stately impression even though he had a thick, short neck and a belly that protruded somewhat; but this was hidden by the good proportions of the rest of his figure. He strode with firm step and held himself like a man; he spoke with a higher voice than one would have expected of someone of his build. He enjoyed good health except for being repeatedly plagued by fevers four years before his death. Toward the end he dragged one foot."

Character and appearance

The strength of Charlemagne's personality was evidently rooted in the unbroken conviction of being at one with the divine will. Without inward contradiction, he was able to combine personal piety with enjoyment of life, a religious sense of mission with a strong will to power, rough manners with a striving for intellectual growth, and intransigence against his enemies with rectitude. In his politically conditioned religiosity, the empire and the church grew into an institutional and spiritual unit. Although his empire survived him by only one generation, it contributed decisively to the eventual reconstitution, in the mind of a western Europe fragmented since the end of the Roman Empire, of a common intellectual, religious, and political inheritance on which later centuries could draw. Charlemagne did not create this inheritance single-handedly, but one would be hard put to imagine it without him. One of the poets at his court called him *rex pater Europae*—"King father of Europe." In truth, there is no other man who similarly left his mark on European history during the centuries of the Middle Ages. (P.Cla.)

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Chaucer

Geoffrey Chaucer, the outstanding English writer before Shakespeare, is among England's greatest poets. He contributed importantly in the second half of the 14th century to the management of public affairs as courtier, diplomat, and civil servant. In that career he was trusted and aided by three successive kings—Edward III, Richard II, and Henry IV. But it is his avocation—the writing of poetry—for which he is remembered. Perhaps the chief characteristics of Chaucer's works are their variety in subject matter, genre, tone, style, and in the complexities presented concerning man's pursuit of a sensible existence. Yet his writings also consistently reflect an all-pervasive humour, combined with serious and tolerant consideration of important philosophical questions. From his writings Chaucer emerges as poet of love, both earthly and divine, whose presentations range from lustful cuckoldry to spiritual union with God. Thereby, they regularly lead the reader to speculation about man's relation both to his fellows and to his Maker, while simultaneously providing delightfully entertaining views of the frailties and follies, as well as the nobility, of mankind.

By courtesy of the trustees of the British Museum

Chaucer, portrait miniature painted after the poet's death. In the British Museum (Harley Ms. 4866).

Forebears and early years. Chaucer's forebears for at least four generations were middle-class English people whose connection with London and the court had steadily increased. John Chaucer, his father, was an important London vintner and a deputy to the king's butler; in 1338 he was a member of Edward III's expedition to Antwerp, in Flanders, now part of Belgium, and he owned property in Ipswich, in the county of Suffolk, and in London. He died in 1366 or 1367 at the age of 53. The name Chaucer is derived from the French word *chaussier,* meaning a maker of footwear. The family's financial success derived from wine and leather.

Although *c.* 1340 is customarily given as Chaucer's birth date, 1342 or 1343 is probably a closer guess. No information exists concerning his early education, although doubtless he would have been as fluent in French as in the Middle English of his time. He also became competent in Latin and Italian. His writings show his close familiarity with many important books of his time and of earlier times.

Chaucer first appears in the records in 1357, as a member of the household of Elizabeth, countess of Ulster, wife of Lionel, third son of Edward III. Geoffrey's father presumably had been able to place him among the group of young men and women serving in that royal household, a customary arrangement whereby families who could do so provided their children with opportunity for the necessary courtly education and connections to advance their careers. By 1359 Chaucer was a member of Edward III's army in France and was captured during the unsuccessful siege of Reims. The King contributed to his ransom, and Chaucer served as messenger from Calais to England during the peace negotiations of 1360. Chaucer does not appear in any contemporary record during 1361–65. He was probably in the King's service, but he may have been studying law—a not unusual preparation for public service, then as now—since a 16th-century report implies that, while so engaged, he was fined for beating a Franciscan friar in a London street. On February 22, 1366, the King of Navarre issued a certificate of safe-conduct for Chaucer, three companions, and their servants to enter Spain. This occasion is the first of a number of diplomatic missions to the continent of Europe over the succeeding 10 years, and the wording of the document suggests that here Chaucer served as "chief of mission."

Capture at the siege of Reims

By 1366 Chaucer had married. Probably his wife was Philippa Pan, who had been in the service of the Countess of Ulster, and who entered the service of Philippa of Hainaut, queen consort of Edward III, when Elizabeth died in 1363. In 1366 Philippa Chaucer received an annuity, and later annuities were frequently paid to her through her husband. These and other facts indicate that Chaucer married well.

In 1367 Chaucer received an annuity for life as yeoman of the king, and in the next year he was listed among the King's esquires. Such officers lived at court and performed staff duties of considerable importance. In 1368 Chaucer was abroad on a diplomatic mission, and in 1369 he was on military service in France. Also in 1369 he and his wife were official mourners for the death of Queen Philippa. Obviously, Chaucer's career was prospering, and his first important poem—*Book of the Duchess*—seems further evidence of his connection with persons in high place.

First important poem

That poem of more than 1,300 lines, probably written in late 1369 or early 1370, is an elegy for Blanche, duchess of Lancaster, John of Gaunt's first wife, who died of plague in September 1369. Chaucer's close relationship with John, which continued through most of his life, may have commenced as early as Christmas 1357 when they, both about the same age, were present at the Countess of Ulster's residence in Yorkshire. For this first of his important poems, Chaucer used the dream-vision form, a genre made popular by the highly influential 13th-century French poem of courtly love, the *Roman de la rose.* Chaucer translated that poem, at least in part, probably as one of his first literary efforts, and he borrowed from it throughout his poetic career. The *Duchess* is also indebted to contemporary French poetry and to Ovid, Chaucer's favourite Roman poet. Nothing in these borrowings, however, will account for his originality in combining dream-vision with elegy and eulogy of Blanche with consolation for John. Also noteworthy here—as it increasingly became in his later poetry—is the tactful and subtle use of a first-person narrator, who both is and is not the poet himself. The device had obvious advantages for the minor courtier delivering such a poem orally before the high-ranking court group. In addition, the *Duchess* foreshadows Chaucer's skill at presenting the rhythms of natural conversation within the confines of Middle English verse and at creating realistic characters within courtly poetic conventions. Also, Chaucer here begins, with the Black Knight's account of his love for Good Fair White, his career as a love poet, examining in late medieval fashion the important philosophic and religious questions concerning the human condition as they relate to both temporal and eternal aspects of love.

Diplomat and civil servant. During the decade of the 1370s, Chaucer was at various times on diplomatic missions in Flanders, France, and Italy. Probably his first Italian journey (December 1372 to May 1373) was for negotiations with the Genoese concerning an English port for their commerce, and with the Florentines concerning loans for Edward III. His next Italian journey occupied May 28 to September 19, 1378, when he was a member of a mission to Milan concerning military matters. Several times during the 1370s, Chaucer and his wife received generous monetary grants from the King and from John of Gaunt. On May 10, 1374, he obtained rent free a dwelling above Aldgate, in London, and on June 8 of that year he was appointed comptroller of the customs and subsidy of wools, skins, and tanned hides for the Port of London. Now, for the first time, Chaucer had a position away from the court, and he and his wife had a home of their own, about a 10-minute walk from his office. In 1375 he was granted two wardships, which paid well, and in 1376 he received a sizable sum from a fine. When Richard II became king in June 1377, he confirmed Chaucer's comptrollership and, later, the annuities granted by Edward III to both Geoffrey and Philippa. Certainly during the 1370s fortune smiled upon the Chaucers.

Appointment as comptroller of the customs

So much responsibility and activity in public matters appears to have left Chaucer little time for writing during this decade. The great literary event for him was that, during his missions to Italy, he encountered the work of Dante, Petrarch, and Boccaccio, which was later to have profound influence upon his own writing. Chaucer's most important work of the 1370s was *Hous of Fame,* a poem of more than 2,000 lines, also in dream-vision form. In some ways it is a failure: it is unfinished, its theme is unclear, and the diversity of its parts seems to overshadow any unity of purpose; but it gives considerable evidence of Chaucer's advancing skill as a poet. The eight-syllable metre is handled with great flexibility; the light, bantering, somewhat ironic tone—later to become one of Chaucer's chief effects—is established; and a wide variety of subject matter is included. Further, the later mastery in creation of memorable characters is here foreshadowed by the marvellous golden eagle who carries the frightened narrator, "Geoffrey," high above the Earth to the houses of Fame and Rumour, so that as a reward for his writing and studying he can learn "tydings" to make into love poems. Here, too, Chaucer's standard picture of his own fictional character emerges: the poet, somewhat dull-witted, dedicated to writing about love but without successful personal experience of it. The comedy of the poem reaches its high point when the pedantic eagle delivers for Geoffrey's edification a learned lecture on the properties of sound. In addition to its comic aspects, however, the poem seems to convey a serious note: like all earthly things, fame is transitory and capricious.

The middle years: political and personal anxieties. In a deed of May 1, 1380, one Cecily Chaumpaigne released Chaucer from legal action: "both of my rape and of any other matter or cause." Rape (*raptus*) could at the time mean either sexual assault or abduction; scholars have not been able to establish which meaning applies here, but, in either case, the release suggests that Chaucer was not guilty as charged. He continued to work at the Customs House and in 1382 was additionally appointed comptroller of the petty customs for wine and other merchandise; but in October 1386 his dwelling in London was leased to another man, and, in December of that year, successors were named for both of his comptrollerships in the customs—whether he resigned or was removed from office is not clear. Between 1382 and 1386 he had arranged for deputies—permanent in two instances and temporary in others—in his work at the customs. In October 1385 he was appointed a justice of the peace for Kent, and in August 1386 he became knight of the shire for Kent, to attend Parliament in October. Further, in 1385 he probably moved to Greenwich, then in Kent, to live. These circumstances suggest that, for some time before 1386, he was planning to move from London and to leave the Customs House. Philippa Chaucer apparently died in 1387; if she had suffered poor health for some time previously, that situation could have influenced a decision to move. On the other hand, political circumstances during this period were not favourable for Chaucer and may have caused his removal. By 1386 a baronial group led by Thomas of Woodstock, duke of Gloucester, had bested both Richard II and John of Gaunt—with whose parties Chaucer had long been associated—and usurped the King's authority and administration. Numerous other officeholders—like Chaucer, appointed by the King—were discharged, and Chaucer may have suffered similarly. Perhaps the best view of the matter is that Chaucer saw which way the political wind was blowing and began early to prepare to move when the necessity arrived.

The period 1386–89 was clearly difficult for Chaucer. Although he was reappointed justice of the peace for 1387, he was not returned to Parliament after 1386. In 1387 he was granted protection for a year to go to Calais, in France, but seems not to have gone, perhaps because of his wife's death. In 1388 a series of suits against him for debts began, and he sold his royal pension for a lump sum. Also, from February 3 to June 4, 1388, the Merciless Parliament, controlled by the barons, caused many leading members of the court party—some of them Chaucer's close friends—to be executed. In May 1389, however, the 23-year-old King Richard II regained control, ousted his enemies, and began appointing his supporters to office. Almost certainly, Chaucer owed his next public office to that political change. On July 12, 1389, he was appointed clerk of the king's works, with executive responsibility for repair and maintenance of royal buildings, such as the Tower of London and Westminster Palace, and with a comfortable salary.

Although political events of the 1380s, from the Peasants' Revolt of 1381 through the Merciless Parliament of 1388, must have kept Chaucer steadily anxious, he produced a sizable body of writings during this decade, some of very high order. Surprisingly, these works do not in any way reflect the tense political scene. Indeed, one is tempted to speculate that during this period Chaucer turned to his reading and writing as escape from the difficulties of his public life. The *Parlement of Foules,* a poem of 699 lines, is a dream-vision for St. Valentine's Day, making use of the myth that each year on that day the birds gathered before the goddess Nature to choose their mates. Beneath its playfully humorous tone, it seems to examine the value of various kinds of love within the context of "common profit" as set forth in the introductory abstract from the *Somnium Scipionis* (*Dream of Scipio*) of Cicero. The narrator searches unsuccessfully for an answer and concludes that he must continue his search in other books. For this poem Chaucer also borrowed extensively from Boccaccio and Dante, but the lively bird debate from which the poem takes its title is for the most part original. The poem has often been taken as connected with events at court, particularly the marriage, in 1382, of Richard II and Anne of Bohemia. But no such connection has ever been firmly established. The *Parlement* is clearly the best of Chaucer's earlier works.

Composition of the *Parlement of Foules*

The Consolation of Philosophy, written by the Roman philosopher Boethius (early 6th century), a Christian, was one of the most influential of medieval books. Its discussion of free will, God's foreknowledge, destiny, fortune, and true and false happiness—in effect, all aspects of the manner in which the right-minded individual should direct his thinking and action to gain eternal salvation—had a deep and lasting effect upon Chaucer's thought and art. His prose translation of the *Consolation* is carefully done, and in his next poem—*Troilus and Criseyde*—the influence of Boethius' book is pervasive. Chaucer took the basic plot for this 8,239-line poem from Boccaccio's *Filostrato.*

Some critics consider *Troilus and Criseyde* Chaucer's finest work, greater even than the far more widely read *Canterbury Tales.* But the two works are so different that comparative evaluation seems fruitless. The state of the surviving manuscripts of *Troilus* shows Chaucer's detailed effort in revising this poem. Against the background of the legendary Trojan War, the love story of Troilus, son

Theme of *Troilus and Criseyde*

of the Trojan king Priam, and Criseyde, widowed daughter of the deserter priest Calkas, is recounted. The poem moves in leisurely fashion, with introspection and much of what would now be called psychological insight dominating many sections. Aided by Criseyde's uncle Pandarus, Troilus and Criseyde are united in love about halfway through the poem; but then she is sent to join her father in the Greek camp outside Troy. Despite her promise to return, she gives her love to the Greek Diomede, and Troilus, left in despair, is killed in the war. These events are interspersed with Boethian discussion of free will and determinism. At the end of the poem, when Troilus' soul rises into the heavens, the folly of complete immersion in sexual love is viewed in relation to the eternal love of God. The effect of the poem is controlled throughout by the direct comments of the narrator, whose sympathy for the lovers—especially for Criseyde—is ever present.

Failure of the *Legend of Good Women*

Also in the 1380s Chaucer produced his fourth and final dream-vision poem, the *Legend of Good Women,* which is not a success. It presents a "Prologue," existing in two versions, and nine stories. In the "Prologue" the god of love is angry because Chaucer had earlier written about so many women who betrayed men. As penance, Chaucer must now write about good women. The "Prologue" is noteworthy for the delightful humour of the narrator's self-mockery and for the passages in praise of books and of the spring. The stories—concerning such women of antiquity as Cleopatra, Dido, and Lucrece—are brief and rather mechanical, with the betrayal of women by wicked men as a regular theme; as a result, the whole becomes more a legend of bad men than of good women. Perhaps the most important fact about the *Legend,* however, is that it shows Chaucer structuring a long poem as a collection of stories within a framework. Seemingly the static nature of the framing device for the *Legend* and the repetitive aspect of the series of stories with a single theme led him to give up this attempt as a poor job. But the failure here must have contributed to his brilliant choice, probably about this same time, of a pilgrimage as the framing device for the stories in *The Canterbury Tales.*

Last years. Chaucer's service as clerk of the king's works lasted only from July 1389 to June 1391. During that tenure he was robbed several times and once beaten, sufficient reason for seeking a change of jobs. In June 1391 he was appointed subforester of the king's park in North Petherton, Somerset, an office that he held until his death. He retained his home in Kent and continued in favour at court, receiving royal grants and gifts during 1393–97. The records show his close relationship during 1395–96 with John of Gaunt's son the Earl of Derby, later King Henry IV. When John died in February 1399, King Richard confiscated John's Lancastrian inheritance; then in May he set forth to crush the Irish revolt. In so doing, he left his country ready to rebel. Henry, exiled in 1398 but now duke of Lancaster, returned to England to claim his rights. The people flocked to him, and he was crowned on September 30, 1399. He confirmed Chaucer's grants from Richard II and in October added an additional generous annuity. In December 1399 Chaucer took a lease on a house in the garden of Westminster Abbey. But on October 25, 1400, he died. He was buried in the Abbey, a signal honour for a commoner.

The Canterbury Tales

Chaucer's great literary accomplishment of the 1390s was *The Canterbury Tales.* In it a group of about 30 pilgrims gather at the Tabard Inn in Southwark, across the Thames from London, and agree to engage in a storytelling contest as they travel on horseback to the shrine of Thomas à Becket in Canterbury, Kent, and back. Harry Bailly, host of the Tabard, serves as master of ceremonies for the contest. The pilgrims are introduced by vivid brief sketches in the "General Prologue." Interspersed between the 24 tales told by the pilgrims are short dramatic scenes presenting lively exchanges, called links and usually involving the host and one or more of the pilgrims. Chaucer did not complete the full plan for his book: the return journey from Canterbury is not included, and some of the pilgrims do not tell stories. Further, the surviving manuscripts leave room for doubt at some points as to Chaucer's intent for arranging the material. The work is, nevertheless, sufficiently complete to be considered a unified book rather than a collection of unfinished fragments. Use of a pilgrimage as a framing device for the collection of stories enabled Chaucer to bring together people from many walks of life: knight, prioress, monk; merchant, man of law, franklin, scholarly clerk; miller, reeve, pardoner; wife of Bath and many others. Also, the pilgrimage and the storytelling contest allowed presentation of a highly varied collection of literary genres: courtly romance, racy fabliau, saint's life, allegorical tale, beast fable, medieval sermon, alchemical account, and, at times, mixtures of these genres. Because of this structure, the sketches, the links, and the tales all fuse as complex presentations of the pilgrims, while, at the same time, the tales present remarkable examples of short stories in verse, plus two expositions in prose. In addition, the pilgrimage, combining a fundamentally religious purpose with its secular aspect of vacation in the spring, made possible extended consideration of the relationship between the pleasures and vices of this world and the spiritual aspirations for the next, that seeming dichotomy with which Chaucer, like Boethius and many other medieval writers, was so steadily concerned.

For this crowning glory of his 30 years of literary composition, Chaucer used his wide and deep study of medieval books of many sorts and his acute observation of daily life at many levels. He also employed his detailed knowledge of medieval astrology and subsidiary sciences as they were thought to influence and dictate human behaviour. Over the whole expanse of this intricate dramatic narrative, he presides as Chaucer the poet, Chaucer the civil servant, and Chaucer the pilgrim: somewhat slow-witted in his pose and always intrigued by human frailty but always questioning the complexity of the human condition, always seeing both the humour and the tragedy in that condition, and always trying to discover the right way for existence on this Earth. At the end, in the "Retractation" with which *The Canterbury Tales* closes, Chaucer as poet and pilgrim states his conclusion that the concern for this world fades into insignificance before the prospect for the next; in view of the admonitions in "The Parson's Tale," he asks forgiveness for his writings that concern "worldly vanities" and remembrance for his translation of the *Consolation* and his other works of morality and religious devotion. On that note he ends his finest work and his career as poet.

Descendants and posthumous reputation. Information concerning Chaucer's children is not fully clear. The probability is that he and Philippa had two sons and two daughters. One son, Thomas Chaucer, who died in 1434, owned large tracts of land and held important offices in the 1420s, including the forestership of North Petherton. He later leased Chaucer's house in Westminster, and his twice-widowed daughter Alice became duchess of Suffolk. In 1391 Chaucer had written *Treatise on the Astrolabe* for "little Lewis," probably his younger son, then 10 years old. Elizabeth "Chaucy," probably the poet's daughter, was a nun at Barking in 1381. A second probable daughter, Agnes Chaucer, was a lady-in-waiting at Henry IV's coronation in 1399. The records lend some support to speculation that John of Gaunt fathered one or more of these children. Chaucer seems to have had no descendants living after the 15th century.

For Chaucer's writings the subsequent record is clearer. His contemporaries praised his artistry, and a "school" of 15th-century Chaucerians imitated his poetry. Over the succeeding centuries, his poems, particularly *The Canterbury Tales,* have been widely read, translated into modern English, and, since about the middle of the 19th century, the number of scholars and critics who devote themselves to the study and teaching of his life and works has steadily increased. (R.M.Lu.)

MAJOR WORKS

Information concerning dates of composition of the extant manuscripts, and of the first printings of the works can be found in the book by Hammond cited in the Bibliography.

LONGER POEMS (in probable order of composition): *Book of the Duchess; Hous of Fame* (unfinished); *Parlement of Foules; Troilus and Criseyde; Legend of Good Women;* prologue, two versions; *The Canterbury Tales;* consisting of The Prologue

(The General Prologue), "The Knight's Tale," "The Miller's Tale," "The Reeve's Tale," "The Cook's Tale," "The Man of Law's Tale," "The Wife of Bath's Tale," "The Friar's Tale," "The Summoner's Tale," "The Clerk's Tale," "The Merchant's Tale," "The Squire's Tale," "The Franklin's Tale," "The Second Nun's Tale," "The Canon's Yeoman's Tale," "The Physician's Tale," "The Pardoner's Tale," "The Shipman's Tale," "The Prioress's Tale," "The Tale of Sir Thopas" and "The Tale of Melibeus" (Chaucer's contributions to the tales told by his fellow-pilgrims, the latter in prose), "The Monk's Tale," "The Nun's Priest's Tale," "The Manciple's Tale," and "The Parson's Tale" (in prose), and ending with Chaucer's "Retractation."

Not all the tales are complete; several contain their own prologues.

SHORTER POEMS: *Anelida and Arcite* (unfinished); *Complaint of Chaucer to his Empty Purse; Lines to Adam Scriven,* his scribe; *Truth, Fortune,* and *Gentilesse;* and letters in verse to Henry Scogan and to Buxton.

PROSE: *The Consolation of Philosophy (Boethius)* (trans. from Boethius' *De consolatione philosophiae); Treatise on the Astrolabe.*

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Editions: The early printed editions of the *Works* are by PYNSON (1526), THYNNE (1532, 1542, and 1545?), STOW (1561), SPEGHT (1598, 1602, and 1687), and URRY (1721); those of the *Canterbury Tales* alone are by CAXTON (*c.* 1478 and *c.* 1484), PYNSON (*c.* 1492), WYNKYN DE WORDE (1495? and 1498), MORELL (1737), and TYRWHITT (1775–78). The most important 19th-century edition is WALTER W. SKEAT, *The Complete Works of Geoffrey Chaucer,* 7 vol. (1894–97). Student editions are FRED N. ROBINSON, *The Works of Geoffrey Chaucer,* 2nd ed. (1957, reissued 1974); ETHELBERT T. DONALDSON, *Chaucer's Poetry: An Anthology for the Modern Reader,* 2nd ed. (1975, reissued 1983); and ALBERT C. BAUGH, *Chaucer's Major Poetry* (1963). Scholarly editions, based on all manuscripts, are JOHN M. MANLY and EDITH RICKERT, *The Text of the Canterbury Tales,* 8 vol. (1940, reissued 1967); and ROBERT K. ROOT, *The Book of Troilus and Criseyde* (1926). Modern editions of the *Tales* include A.C. CAWLEY, *The Canterbury Tales* ("Everyman's Library," 1958, reissued 1975); and ROBERT A. PRATT, *Selections from the Tales of Canterbury, and Short Poems* ("Riverside Editions," 1966).

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(R.M.Lu./Ed.)

Chemical Compounds

A chemical compound is a substance made up of identical molecules containing atoms of two or more elements. Millions of chemical compounds are known; millions more probably exist in nature; and still other millions may be synthesized in the laboratory. The known compounds differ from one another in almost every conceivable way. Some, such as water, are colourless; others, the dyes and pigments, embody every colour of the rainbow. Some are light gases at ordinary temperatures; others, dense solids. Some are chemically inert; others, so unstable that they exist for only fractions of a second. In order to deal with the bewildering mass of diverse information about this multitude of substances, chemists have devised a number of classification schemes. These provide rational ways of organizing all chemical compounds, known or still to be discovered, into related groups of similar and dissimilar substances.

In any systematic classification, of course, chemical compounds must themselves be distinguished from other forms of matter. Most matter encountered in everyday life consists of mixtures of two or more substances. Generally, such mixtures can be separated by physical means—as, for example, salt is separated from seawater when the water is evaporated. The pure substances themselves are either atoms, or aggregates of atoms, of elements or molecules of compounds. Because the molecules of compounds contain atoms of several elements, compounds can be broken down into their elements by chemical reactions.

Molecular formulas represent the compositions of compounds in terms of the specific elements that constitute them—that is, in terms of what is called their elemental composition. Such formulas consist of the chemical symbols of the elements involved, along with numerical subscripts that indicate the numbers of atoms of each kind in the molecules. Thus, the formula for water, H_2O, indicates that a molecule of water contains two atoms of hydrogen and one of oxygen.

Among the most useful schemes for the classification of chemical compounds are those based on (1) the elemental composition of the compound, (2) the types of bonds that exist between the atoms in their molecules, and (3) the varieties of chemical reactions they undergo. Several other, less important classification methods are used for certain special purposes. (W.F.K.)

Organic compounds

Classification by elemental composition. *Organic compounds.* Early in the 19th century, chemists began to classify chemical compounds found in or produced by living organisms as organic. By contrast, compounds from nonliving sources, such as minerals, were classified as inorganic. Like the biological materials from which they come, organic compounds, such as sugar or alcohol, either burn or char when heated in air; inorganic compounds, such as limestone or table salt, either are inert toward heating or undergo specific characteristic transformations not associated with decomposition. Although it has been known since 1828 that typical organic compounds can be synthesized in the laboratory from inorganic compounds without the involvement of living organisms, the name organic still is used to refer to this group of compounds. Generally, organic compounds are compounds of carbon—except for simple compounds containing only carbon and oxygen and minerals, such as limestone or marble. Compounds involving elements other than carbon are classed as inorganic.

Aliphatic and aromatic compounds

Though carbon is only one element among more than 100, it is thought that the number of carbon compounds is larger than the number of compounds of all other elements put together. Carbon atoms form strong bonds with the atoms of nonmetals, such as hydrogen, oxygen, nitrogen, and chlorine, but the large number of carbon compounds is due to the fact that carbon atoms also are able to form strong bonds with other carbon atoms. The atoms of no other element bond to one another to the same extent or in the same variety of ways as carbon atoms do. Carbon atoms join together in chains or rings that function as molecular backbones to which atoms of other elements are attached. These chains can vary in length from two carbon atoms, as in ethane (symbolized by the formula C_2H_6), a component of natural gas, to hundreds of thousands of carbon atoms, as in polyethylene, a manufactured polymer used to make household utensils and fibres. Each carbon atom in any chain has attached to it a variety of atoms or groups of atoms, which may also be other carbon chains. Molecules consisting of simple carbon chains have been called aliphatic ever since the early discovery that they occurred in fatty substances (Greek *aleiphar-*, "fat"). In contrast, certain compounds consisting of chains of carbon atoms whose ends are joined to form rings, of which a prototype molecule is benzene, C_6H_6, are called aromatic compounds, a name originally chosen because of the odour associated with coal tar, in which such molecules typically are found. The discovery that benzene is a ring of six carbon atoms, rather than a chain, was one of the dramatic breakthroughs in 19th-century science. Generally, aliphatic and aromatic compounds behave quite differently, and the subclassification of organic compounds into these two large families, aliphatic and aromatic, is a natural one. Ring compounds are also called cyclic compounds.

Organic compounds often contain special groups of atoms called functional groups, which readily undergo chemical reaction. Because these functional groups dominate the chemistry of the compounds that contain them, they provide a convenient way of classifying organic compounds. Alcohols, ethers, aldehydes, ketones, and carboxylic acids are typical examples of organic compounds classified by the functional groups they contain.

Inorganic compounds. The periodic table provides the chemist with his most comprehensive scheme for classification of the elements. Consequently, it also serves as a starting point for the classification of inorganic compounds, which, as mentioned above, are compounds of elements other than carbon. One convenient way to classify inorganic compounds is by grouping together all of the compounds of a particular element. This is especially useful with compounds containing atoms of only two elements (binary compounds), which are designated commonly by the suffix -ide. Thus, the compound sodium chloride (symbolized by the formula NaCl), or table salt, is formed by the combination of the metallic element sodium (Na) with the nonmetallic element chlorine (Cl). Chlorides of other metallic elements are known, and the definition of chlorides as a group is useful in certain classification schemes. This type of classification can easily be extended to other elements of the chlorine family in the periodic table, which form compounds similar to the chlorides. Members of this family are known as halogens (salt formers). Consequently, the binary compounds formed by this family of elements are called by the generic name halides. Familiar examples of halides are sodium fluoride (NaF), the material added to drinking water to prevent tooth decay, and silver iodide (AgI), a substance used in making photographic films.

Hydrides, oxides, and sulfides

Other examples of classifying together the binary compounds of a single element are the binary compounds of hydrogen, known as hydrides, those of oxygen, called oxides; and those of sulfur, called sulfides. Like the halogen elements, most elements within a single vertical file, or group, of the periodic table form similar compounds. For example, all of the elements of the family known as the alkali metals form similar chlorine compounds, which are known collectively as the alkali chlorides. Similarly, the elements of the family known as the alkaline-earth metals form oxygen compounds known as the alkaline-earth oxides.

Similar principles apply to the classification of inorganic compounds composed of more than two elements. In naming these compounds, suffixes and prefixes are employed along with roots. The suffix -ate, for example, often is used to indicate that an element contains a number of oxygen atoms bound to it. Thus, sodium sulfate is the name of the compound with the formula Na_2SO_4 (consisting of sodium, sulfur, and oxygen), and calcium silicate is $CaSiO_3$ (calcium, silicon, and oxygen). Sulfates and silicates are recognizable classes of compounds, and, again, classes such as the alkali sulfates and the alkaline-earth silicates often are discussed.

Hemoglobin and chlorophyll

Organometallic compounds. The great number and variety of potential combinations among the atoms of the elements make it impossible for any classification scheme to put all possible compounds into mutually exclusive categories. Thus, it is found that many metal atoms, most frequently encountered in inorganic substances, can also combine with compounds generally classed as organic. The resulting organometallic, or metallo-organic, compounds are widespread among substances essential to living systems. Hemoglobin, the vital oxygen-carrying constituent in the blood of higher animals, is a large, carbon-containing molecule in which an iron atom is combined. Chlorophyll, similarly essential to green plants, has a magnesium atom imbedded in a large organic molecule. Vitamin B_{12} is an organometallic compound containing cobalt. But organometallics are not restricted to compounds found in nature. Many new synthetic varieties are produced each year. The compound tetraethyllead, formula $Pb(C_2H_5)_4$ (one atom of lead bonded to four groups, each consisting of two atoms of carbon and five of hydrogen), is a synthetic organometallic compound that has found wide use as an additive to gasoline.

Variation in ratios of elements

Nonstoichiometric compounds. One of the tenets of the atomic theory, on which much of chemistry is based, is that, when elements combine, they do so in such a manner that the ratio of the numbers of atoms of elements is integral. This rule is implied in typical molecular formulas. For example, the formula Al_2O_3 for the substance aluminum oxide means that atoms of aluminum combine with atoms of oxygen in the exact ratio of 2:3 to form this compound. The ratio of elements in a compound is revealed by the analysis of any pure sample of it. Any compound found to contain atoms of elements in whole-number ratios is said to be a stoichiometric (Greek: "element measuring") compound. A great majority of compounds, including nearly all organic compounds, fall into this class. There is, however, another class of compounds, nonstoichiometric compounds, in which the relative numbers of elemental atoms are not integral. The class includes, particularly, the oxides and sulfides of certain metals. For example, accurate analyses have shown that the number of zinc atoms relative to sulfur atoms in zinc sulfide, ZnS, can actually exceed the predicted 1:1 ratio of the formula by 0.03 percent. Similarly, the analysis of a sulfide of cerium suggests a formula $Ce_{2{\cdot}7}S_4$, which means that the ratio of cerium to sulfur is in the proportion of 2.7:4. The electricity-conducting properties of such nonstoichiometric substances make them increasingly important as components of solid-state devices used in the electronics industry.

Classification by bond type. *Ionic compounds.* In the chemical reaction by which sodium chloride is formed from sodium atoms and chlorine atoms, the sodium atoms each lose an electron (the basic, negatively charged particle present in all atoms and also free in space), and the chlorine atoms each gain one. The consequence of this kind of electron transfer is that the compound formed exists as an aggregate of particles of matter bearing electrical charges, or ions. Forces of electrical attraction hold, or bond, the positively charged sodium ions in a matrix of negatively charged chloride ions, and vice versa. This form of bonding, ionic bonding, is characteristic of all salts. Certain hydrides (*e.g.,* sodium hydride) and oxides (*e.g.,* magnesium oxide) also have properties—such as high melting point and electrical conductivity in solution—that classify them as saltlike, or ionic, compounds.

Covalent compounds. An alternative, nonionic process, however, is involved in the combination of certain atoms to form compounds. In this process, electrons are shared between atoms, producing bonds, which are referred to as covalent bonds, between them. Covalent bonding occurs when carbon atoms combine in organic compounds and between nonmetal atoms when they combine into molecules. Such bonding results in the formation of covalent molecules, typical examples of which are methane (CH_4), a constituent of natural gas, and water (H_2O). Covalent molecules are complete entities, not electrically charged fragments like ions, and their aggregation into large samples is the consequence of the attractive forces between molecules, rather than electrostatic attraction. Generally speaking, intermolecular forces are weaker than electrostatic forces. As a result, the molecules of covalent compounds are pulled apart more easily than ions are and therefore the melting points of covalent compounds are lower than those of ionic compounds.

Macro-
molecules

The mechanism of covalent-bond formation leads in some cases to the formation of huge, extended molecules involving thousands to millions of atoms. These macromolecules (macro-, "large") are also referred to as polymers (poly-, "many"; -mers, "parts"). The familiar plastics encountered in thousands of everyday uses are polymers. Polyethylene, Plexiglas, and nylon are examples. These are huge organic molecules formed by the combination of many small units. Varieties of macromolecules important to living organisms are proteins, carbohydrates, and nucleic acids, such as deoxyribonucleic acid (DNA). All of these are polymers held together by covalent bonds. There are also some inorganic materials that properly can be classed as covalent macromolecules. Quartz (SiO_2) and other silicon-containing minerals, such as clays, have many chemically identical small units bonded together in two- and three-dimensional networks.

Coordination compounds. There are many stable compounds the structures of which are such that they can form even more stable compounds by adding other molecules. Such addition compounds are called coordination compounds. An example is expressed by the equation

$$CuSO_4 + 4NH_3 \rightarrow Cu(NH_3)_4SO_4,$$

which indicates that one molecule of cupric sulfate ($CuSO_4$), also written copper(II) sulfate, and four molecules of ammonia (NH_3) combine to produce the coordination compound tetramminecopper(II) sulfate. In the process, four covalent bonds are formed between the four ammonia molecules and the copper ion. In each of these bonds, a pair of electrons on the ammonia molecule is shared with the copper ion. In such bonds, the electrons are donated by only one of the two atoms, rather than coming from both atoms, as in typical covalent bonds. A great variety of coordination compounds is formed by the so-called transition elements—*i.e.,* metallic elements found in the central portion of the periodic table. An important example of a coordination compound is the one formed when carbon monoxide attaches to the iron atom in hemoglobin of the blood. The stability of this compound robs hemoglobin of its oxygen-carrying ability. For an organism in which a major percentage of the hemoglobin molecules have combined with carbon monoxide, the consequence is death.

Donation
of electrons

Classification by chemical reactivity. A natural consequence of the accumulation of information about the behaviour of chemical compounds is the recognition of similarities in their chemical properties. This recognition leads to classification on the basis of chemical reactivity, a classification that generally correlates well with classifications based on chemical structure.

Acids, bases, and salts. Very early in the development of chemistry, it was recognized that a type of reaction common to many substances is their ability to act as acids. Vinegar, for example, is a natural product that behaves as it does because it contains an organic acid, acetic acid. Acids generally have the property of imparting a characteristic colour to certain dyestuffs used as indicators. The indicator litmus, for instance, turns from blue to pink in acidic solutions. Acids also have a corrosive effect on many metals and produce hydrogen gas as a result of the action. This observation suggests that acids are hydrogen-containing compounds, as, indeed, most of them are. The so-called mineral acids are aqueous solutions of compounds derived from inorganic materials; typical mineral acids are hydrochloric acid (HCl), sulfuric acid (H_2SO_4), and nitric acid (HNO_3).

Bases, on the other hand, are a class of compounds capable of reacting with acids to neutralize their acidity. They, too, impart a characteristic colour to indicator dyes; litmus turns from pink to blue in solutions of bases. Alkali hydroxides—*e.g.,* sodium hydroxide, (NaOH), commonly called lye—and alkaline-earth oxides—*e.g.,* calcium oxide (CaO), familiarly called lime—are typical bases. Two neutralization reactions are shown in the following equations:

$$HCl + NaOH \rightarrow HOH + NaCl;$$

$$H_2SO_4 + CaO \rightarrow HOH + CaSO_4.$$

In the first equation, hydrochloric acid is neutralized with sodium hydroxide to give water and sodium chloride, common table salt. In the second reaction, sulfuric acid comes together with calcium oxide to form, again, water and a salt, in this case, calcium sulfate. From these reactions and many others like them, it can be generalized that acids as a class react with bases as a class to produce water and a third class of substances, salts.

Some substances, however—ammonia, for example—are capable of neutralizing acids without giving water as a product. The reaction of ammonia with nitric acid can be written

$$NH_3 + HNO_3 \rightarrow NH_4NO_3.$$

In this reaction, a molecule of ammonia is seen to combine with a hydrogen ion (H^+) from the acid in a manner similar to the way in which oxygen (or an oxygen–hydrogen group) combines with hydrogen in the above examples. One definition of bases, then, calls them the class of compounds capable of accepting hydrogen from acids. Ammonium nitrate (NH_4NO_3) is another example of a salt, the product of the neutralization of an acid by a base.

To enlarge the classification, there are compounds containing no hydrogen that are also capable of reacting with bases. Boron trifluoride (BF_3) is such a compound. It reacts with ammonia to produce a neutral compound, much like a salt: $NH_3 + BF_3 \rightarrow H_3NBF_3$. The structural interpretation of this reaction is that the boron atom accepts a share in a pair of electrons from the nitrogen atom of ammonia. A covalent bond is formed between boron and nitrogen. This reaction suggests another definition of acids and bases: acids are compounds capable of accepting a pair of electrons; bases are compounds capable of donating a pair of electrons.

Acids as
electron
recipients

Oxidizing and reducing agents. Another frequently employed classification of chemical compounds is based upon their behaviour in so-called oxidation–reduction reactions. Such reactions are ones in which the combining capacity (valence) of several of the components is changed by the reaction. A simple example is:

$2Cu$ (copper) $+ O_2$ (oxygen) $\rightarrow 2CuO$ (copper[II] oxide).

In this reaction, two atoms of copper are said to be oxidized to two molecules of copper(II) oxide by a molecule of oxygen. In the process, the valence of the copper atoms has been changed from 0 in the uncombined element to +2 in the compound. At the same time, the valence of oxygen has been reduced from 0 in the elemental form to −2 in the compound. A comparable result can be accomplished by treating copper with sulfuric acid:

$$Cu + H_2SO_4 \rightarrow CuO + SO_2 + H_2O.$$

In this reaction, copper again is oxidized; and sulfuric acid is the oxidizing agent, being reduced to sulfur dioxide in the process. (The valence of sulfur changes from +6 in sulfuric acid to +4 in sulfur dioxide.) There is a complementarity in this classification scheme: in oxidation–reduction reactions, oxidizing agents always are reduced; reducing agents invariably are oxidized.

Because examples of oxidation and reduction are so numerous in all branches of chemistry, the classification of substances as oxidizing or reducing agents is important. The results of many reactions can be anticipated by recognizing the degree to which the compounds involved exhibit oxidizing or reducing tendencies.

Nonchemical (trivial) classification schemes. It is hardly surprising that, in any study as vast as the examination of all kinds of matter, there are a great many classification schemes based simply on convenience. Thus, for example, classification based upon the physical state of a pure substance divides all substances into the categories gases, liquids, and solids. Although such a classification may be useful, it reflects little about the essential structure or reactivity of a material. For example, the familiar forms of the compound H_2O are solid ice, liquid water, and gaseous steam. The compound itself, however, is chemically identical in all three forms.

Natural and synthetic compounds

Another nonchemical basis for a classification scheme may be the source of a compound. Thus, natural compounds are found in nature; synthetic compounds are produced in laboratories or factories. Natural compounds may be further classified as terrestrial, meteoric, or lunar, depending on their place of origin. Again, however, it is important to recognize that a pure compound isolated from any one source is chemically identical with the same compound from any other source. It is in this sense that the chemist considers all such nonchemical classification schemes as trivial. (W.F.K.)

INORGANIC CHEMICAL COMPOUNDS

In the early stages of the development of chemistry, considerable emphasis was placed on the origins of the various substances. As noted earlier, those substances clearly derived from living organisms were referred to as organic, whereas those that had in no known way been associated with living organisms were described as inorganic. Since the common substances derived from living matter are composed predominantly of the elements carbon, nitrogen, oxygen, and hydrogen—especially carbon—much of the chemistry of those elements was reserved for the field of organic chemistry, organic compounds being chiefly those substances in which varying amounts of hydrogen, oxygen, or nitrogen are combined with carbon. Inorganic chemistry, then, encompasses the bulk of the chemistry of all the remaining elements; see Figure 1. In addition, there exist compounds of carbon, hydrogen, oxygen, and nitrogen that are strictly inorganic, and there are many inorganic compounds formed by combination of various of these elements with other known elements.

period \ group	1 Ia	2 IIa	3 IIIb	4 IVb	5 Vb	6 VIb	7 VIIb	8	9 VIII	10	11 Ib	12 IIb	13 IIIa	14 IVa	15 Va	16 VIa	17 VIIa	18 0
1	1 H																1 H	2 He
2	3 Li	4 Be											5 B	6 C	7 N	8 O	9 F	10 Ne
3	11 Na	12 Mg											13 Al	14 Si	15 P	16 S	17 Cl	18 Ar
4	19 K	20 Ca	21 Sc	22 Ti	23 V	24 Cr	25 Mn	26 Fe	27 Co	28 Ni	29 Cu	30 Zn	31 Ga	32 Ge	33 As	34 Se	35 Br	36 Kr
5	37 Rb	38 Sr	39 Y	40 Zr	41 Nb	42 Mo	43 Tc	44 Ru	45 Rh	46 Pd	47 Ag	48 Cd	49 In	50 Sn	51 Sb	52 Te	53 I	54 Xe
6	55 Cs	56 Ba	57 La *	72 Hf	73 Ta	74 W	75 Re	76 Os	77 Ir	78 Pt	79 Au	80 Hg	81 Tl	82 Pb	83 Bi	84 Po	85 At	86 Rn
7	87 Fr	88 Ra	89 Ac **	104 Unq	105 Unp	106 Unh	107 Uns	108 Uno	109 Une	110 Uun								

* 6	58 Ce	59 Pr	60 Nd	61 Pm	62 Sm	63 Eu	64 Gd	65 Tb	66 Dy	67 Ho	68 Er	69 Tm	70 Yb	71 Lu
** 7	90 Th	91 Pa	92 U	93 Np	94 Pu	95 Am	96 Cm	97 Bk	98 Cf	99 Es	100 Fm	101 Md	102 No	103 Lr

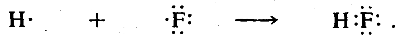
Figure 1: The periodic table of the elements.

CHEMICAL BONDING IN INORGANIC COMPOUNDS

The atoms of the elements that are combined in an inorganic compound may be bound to each other by a variety of kinds of chemical bonds, as the attractive forces that hold atoms together are called. These are most broadly separable into three categories: covalent bonds, electrovalent bonds, and metallic bonds. Within certain of these categories, there are also a few other bond types that must also be recognized.

Sharing of electron pairs

Covalent bonding. A covalent bond involves the binding together of two atoms of either the same element or of two different elements by the sharing of two electrons (an electron pair). This type of bond is illustrated by the formation of the compound hydrogen fluoride from its constituent atoms of hydrogen and fluorine, as shown in the following chemical equation:



$$H\cdot \quad + \quad \cdot\ddot{\underset{\cdot\cdot}{F}}: \quad \longrightarrow \quad H:\ddot{\underset{\cdot\cdot}{F}}:.$$

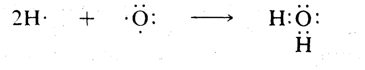
hydrogen atom — fluorine atom — hydrogen fluoride molecule

where atoms of hydrogen and fluorine are represented (as is customary) by their chemical symbols, H and F, respectively; the dots represent electrons in the outermost shells of the atoms of the elements (every atom being made up of electrons in concentric shells around a central nucleus). As is generally true in chemical equations, the starting materials are written to the left of an arrow (which points in the direction of the reaction) and the products to the right.

Since every atom has a definite electron capacity in its outermost shell of electrons (also called the valence shell), each of the combining atoms must be able to accommodate the shared electrons within its valence shell. Very commonly, the number of electrons that can be accommodated is eight. This is the so-called rule of eight (or octet rule), and it comes about because the two lowest energy subshells in any major electron shell are the *s* (electron capacity of two) and *p* (electron capacity, six) subshells (with the letters serving only to designate the subshells and having no other pertinent meaning).

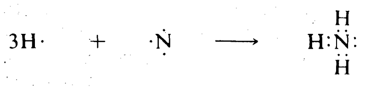
The rule of eight

The formulas of the covalent compounds water (H_2O, O being an atom of oxygen) and ammonia (NH_3, N being an atom of nitrogen) reflect the rule of eight (as does that of hydrogen fluoride, above). This circumstance is shown in the equations below, in which the numbers of electrons in the free atoms reflect the electronic structures of the free elements, whereas the number of electrons assigned



$$2H\cdot \quad + \quad \cdot\ddot{\underset{\cdot}{O}}: \quad \longrightarrow \quad H:\underset{H}{\ddot{\underset{\cdot\cdot}{O}}}:$$

hydrogen atoms — oxygen atom — water molecule



$$3H\cdot \quad + \quad \cdot\dot{\underset{\cdot}{N}}: \quad \longrightarrow \quad H:\underset{H}{\overset{H}{\dot{\underset{\cdot\cdot}{N}}}}:$$

hydrogen atoms — nitrogen atom — ammonia molecule to each element in the compounds includes those not involved in bonds as well as those utilized by that element in covalent bonding.

Electron configuration. A free hydrogen atom has the electronic configuration designated $1s^1$, in which the first number indicates the number of the shell proceeding outward, the letter designates the type of subshell, and the superscript represents the number of electrons involved. The valence-shell configuration for oxygen is $2s^22p^4$ and that for nitrogen is $2s^22p^3$. In its covalent compounds, hydrogen effectively has the configuration $1s^2$, whereas the configurations of both oxygen and nitrogen in covalent compounds are $2s^22p^6$.

The closed-shell configuration of hydrogen

The closed-shell configuration of bonded hydrogen ($1s^2$) thus contains only two electrons, rather than eight; there is one important respect, however, in which this closed-shell configuration closely resembles those of such other covalently bonded elements as oxygen and nitrogen in their compounds water and ammonia ($2s^22p^6$). In each case, the bonded element has gained a share in a sufficient number of electrons to have the same configuration as the nearest noble gas (or Group 0 element) in the periodic table (see Figure 1). With hydrogen, this element is helium (configuration $1s^2$), whereas with oxygen and nitrogen it is neon ($2s^22p^6$). One may conclude that when an element obeys the rule of eight, it has assumed a noble gas configuration.

For elements having atomic numbers greater than 10 (the elements customarily being arranged in order of increasing atomic number—a value equivalent to the number of electrons in the neutral atom), the valence shell contains a d subshell in addition to the s and p subshells. Many compounds of these elements reflect the fact that more than eight electrons can be accommodated in the outermost shells.

Electronegativity. The ability of two atoms to share electrons depends on their having comparable relative abilities to attract electrons. This affinity for electrons depends ultimately on the charge on the nucleus of the atom as moderated by the other electrons in the atom. It is commonly estimated by a quantity called the electronegativity, which increases from left to right and from bottom to top in the periodic table (Figure 1). If two covalently bonded atoms have identical electronegativities, they share the electron pair equally. The bond is then said to be nonpolar, and the molecule is described as being homopolar (or nonpolar). If the electronegativities of the two atoms are not equal, the electron pair of the covalent bond is shared unequally. This causes one of the atoms to have a slight excess of negative charge and the other a slight excess of positive charge. Such a bond is called polar, and the molecule is described as being heteropolar (or simply polar). Polar and nonpolar molecules have different properties.

Polarity

Formation of ions

Electrovalent bonding. Electrovalent bonding occurs when the electronegativity difference between two atoms (elements) is so great that electron sharing is not possible. In this case, one atom takes electrons away from the other, and the two atoms are converted to charged particles called ions. The resulting ions have opposite charges, and they are attracted to each other by the resulting electrostatic force (positive–negative attraction). An example of how ions can be considered to be formed from neutral atoms is shown below in the reaction of sodium with fluorine:

$$\begin{array}{ccccccc} \mathrm{Na}\cdot & + & \cdot\overset{\cdot\cdot}{\underset{\cdot\cdot}{\mathrm{F}}}: & \longrightarrow & \mathrm{Na}^{+} & + & :\overset{\cdot\cdot}{\underset{\cdot\cdot}{\mathrm{F}}}:^{-}. \\ (3s^1) & & (2s^22p^5) & & (2s^22p^6) & & (2s^22p^6) \\ \text{sodium atom} & & \text{fluorine atom} & & \text{sodium ion} & & \text{fluoride ion} \end{array}$$

A great difference between such ion–ion bonds and ordinary covalent (non-ionic) bonds is the fact that the electrostatic bond is not restricted to a single pair of atoms. Every negative ion attracts every other positive ion in its vicinity, and vice versa. This generalized attraction results in the formation of ionic crystal lattices (extended structures built up of positive and negative ions) by solid electrovalent compounds. For an example see Figure 2, representing the structure of sodium chloride (NaCl) crystal.

The number of electrons transferred between atoms is usually in accord with the rule of eight, as illustrated in the examples below:

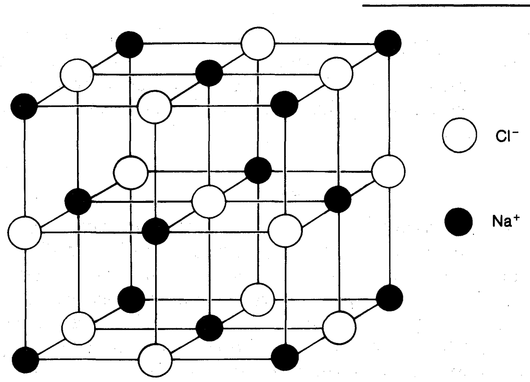


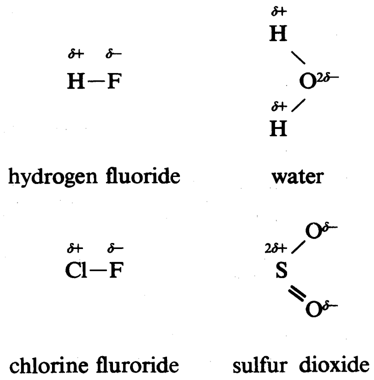
Figure 2: Ionic structure of sodium chloride in a three-dimensional array.

$$\begin{array}{ccccccc} \cdot\mathrm{Ca}\cdot & + & \cdot\overset{\cdot\cdot}{\underset{\cdot}{\mathrm{O}}}: & \longrightarrow & \mathrm{Ca}^{2+} & + & :\overset{\cdot\cdot}{\underset{\cdot\cdot}{\mathrm{O}}}:^{2-} \\ (4s^2) & & (2s^22p^4) & & (3s^23p^6) & & (2s^22p^6) \\ \text{calcium atom} & & \text{oxygen atom} & & \text{calcium ion} & & \text{oxygen ion} \end{array}$$

$$\begin{array}{ccccccc} 3\mathrm{Li}\cdot & + & \cdot\overset{\cdot\cdot}{\underset{\cdot}{\mathrm{N}}}\cdot & \longrightarrow & 3\mathrm{Li}^{+} & + & :\overset{\cdot\cdot}{\underset{\cdot\cdot}{\mathrm{N}}}:^{3-}. \\ (2s^1) & & (2s^22p^3) & & (2s^2) & & (2s^22p^6) \\ \text{lithium atom} & & \text{nitrogen atom} & & \text{lithium ion} & & \text{nitrogen ion} \end{array}$$

Polar molecules, such as those of water, bind to ions by attractive forces between the partially separated charges (dipoles) of the polar molecules and the free charges of the ions. When sodium chloride and other salts dissolve in water, their ions interact with the molecules of solvent by means of ion–dipole forces, as shown in Figure 3 (Left).

Molecules that have unsymmetrical electronic arrangements (dipoles) are commonly referred to as polar because their extremities bear opposite charges, which are indicated by the symbols $\delta+$ and $\delta-$. Various examples of polar molecules and their respective dipoles are provided below:



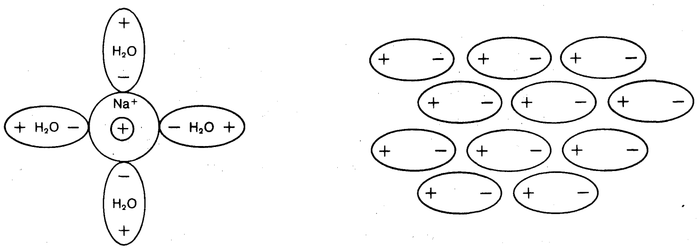


Figure 3: (Left) Attraction of water molecules to a sodium cation by ion–dipole forces. (Right) Alignment of dipoles in a liquid or solid phase containing polar molecules.

Polar molecules attract each other by aligning their charges so that those of opposite sign are adjacent. Dipole–dipole forces provide the intermolecular attractions that hold the molecules of many solids and liquids together, as shown in Figure 3 (Right). Certain molecules with polar bonds do not have dipole moments because of the cancelling of dipoles. Carbon dioxide, CO_2, provides an example:

$$\overset{\delta-}{:\ddot{O}} = \overset{2\delta+}{C} = \overset{\delta-}{\ddot{O}:}\ .$$

Hydrogen bonds

Polar molecules with hydrogen atoms at their positive extremities often form very strong ion–dipole and dipole–dipole bonds. This is particularly true when the second element in the compound is strongly electronegative, as are fluorine, oxygen, and nitrogen. These linkages are called hydrogen bonds and are responsible for the unusually large mutual attractions of the molecules of such compounds as hydrogen fluoride, water, and ammonia (when compared to their higher molecular weight analogues: hydrogen chloride, hydrogen sulfide, and phosphine, respectively). Comparisons of the heat of vaporization (the amount of heat needed to vaporize a standard quantity) of the hydrogen halides and of water and its heavier analogues show that more heat is needed to vaporize the lightest member of each series, and hence the strength of the dipole–dipole attractions is markedly greater for these compounds (more heat being required to overcome stronger attractions).

The nature of the forces holding the particles together in solids determines the physical and chemical properties of those substances, as shown in Table 1. Ion–ion forces and metallic bonding (the attractive force between atoms in metals) are usually extremely strong. Ion–dipole and dipole–dipole forces are not so strong, and the so-called van der Waals forces (weak attractions between nonpolar molecules) are weakest of all.

More subtle variations in properties than are implied in Table 1 are illustrated by the melting points of the fluorides for the horizontal series (period) of elements in the periodic table, beginning with sodium: 988° C (1,810° F), sodium fluoride; 1,266° C (2,311° F), magnesium fluoride; 1,291° C (2,356° F), aluminum fluoride; −90.2° C (−130.4° F), silicon tetrafluoride; −83° C (−117° F), phosphorus pentafluoride; −50.5° C (−58.9° F), sulfur hexafluoride. In this series of compounds, the covalent nature of the molecular structures increases in proceeding to the right.

Table 1: Binding Forces and Properties of Solids

class of solid	nature of bonds	physical properties	example
Ionic	ion–ion (electrovalent)	hard, brittle, high mp and bp, melt conducts electricity	$NaCl$, CaF_2, LaF_3
Giant molecular	covalent	very hard, high mp, and bp, nonconductor	diamond, boron nitride, carborundum
Polar molecular	dipole–dipole	moderately soft, intermediate mp, nonconductor	water, ammonia, ether
Metallic	metallic bonds	great variety in hardness, mp, and bp, malleable, ductile, electrical conductors	elemental metals; *e.g.*, Ag, Cu, Fe, Na, Hg, W
Nonpolar molecular	van der Waals	very soft, low mp and bp, nonconductor	naphthalene, CCl_4, N_2, Ar

NOMENCLATURE OF INORGANIC COMPOUNDS

Binary compounds. There are several acceptable ways of naming binary compounds (compounds of only two elements). The choice of which is used depends on the complexity of the chemistry of the elements in the particular compound and the complexity of the structure of the compound itself.

Order of names

For compounds of electropositive elements (that is, those with only positive oxidation states) that do not have variable valences, such as the alkali and alkaline-earth elements, the name is composed of the unmodified name of the electropositive element followed by that of the electronegative element (the element with negative oxidation state), the latter having its name modified in a phonetically useful way to end in "-ide." The simple compound of sodium and chlorine (NaCl), for example, is called sodium chloride; that between calcium and sulfur (CaS) is calcium sulfide; that between magnesium and nitrogen (Mg_3N_2) is magnesium nitride; and the compound formed between aluminum and boron (AlB_{12}) is known as aluminum boride.

Names incorporating oxidation states

In cases in which several oxidation states are possible for the electropositive element in a binary compound, a roman numeral is added in parentheses after the name of the electropositive element. Thus, the compound with the molecular formula $FeCl_2$ is known as iron(II) chloride, whereas the compound with the formula $FeCl_3$ is referred to as iron(III) chloride. Similarly, CuCl is copper(I) chloride; $SnBr_4$ is tin(IV) bromide; NO_2 is nitrogen(IV) oxide; and SF_6 is sulfur(VI) fluoride. This system replaces an older system in which electropositive elements were given different suffixes to indicate the oxidation state. In the older system, for example, iron(II) chloride was called ferrous chloride and iron(III) chloride was known as ferric chloride.

Alternatively, when compounds incorporating different combinations of the same two elements are known, Latin and Greek prefixes are used to indicate the proportions of elements present in a given compound. A series of compounds named by this method are given in Table 2.

Table 2: Some Compounds Named Using Prefixes To Indicate Proportions

molecular formula	name
N_2O	dinitrogen oxide
N_2O_3	dinitrogen trioxide
N_2O_4	dinitrogen tetroxide
N_2S_5	dinitrogen pentasulfide
S_2Cl_2	disulfur dichloride
Al_2Cl_6	dialuminum hexachloride
ClO_3	chlorine trioxide

Names indicating ratios

This system is also often used more loosely to indicate the ratios of elements combined in a substance rather than its exact molecular formula. Thus the name aluminum trichloride (and the formula $AlCl_3$) indicates the composition of the substance but not its true molecular formula (which is generally thought to be Al_2Cl_6). Furthermore, manganese dioxide, MnO_2, exists in a saltlike lattice (see below *Structural classification of inorganic compounds*) rather than as discrete molecules as the name suggests.

The binary acids may either be named as typical binary compounds (hydrogen chloride for HCl, hydrogen sulfide for H_2S) or they may be given the prefix hydro- and the suffix -ic acid (thus hydrochloric acid for HCl, and hydrosulfuric acid for H_2S).

Certain classes of binary compounds have systems of nomenclature based on common names. This is especially true of hydrogen derivatives, both neutral and charged, as shown in Table 3.

Table 3: Common Names of Hydrogen Compounds

neutral compound		cation		anion	
formula	name	formula	name	formula	name
H_2O	water	H_3O^+	hydronium	OH^-	hydroxide
NH_3	ammonia	NH_4^+	ammonium	NH_2^-	amide
N_2H_4	hydrazine	$N_2H_5^+$	hydrazinium	—	—
B_2H_6	diborane	—	—	BH_4^-	tetrahydridoborate
SiH_4	silane	—	—	—	—

Compounds with common names

The hydrides of boron are named as boranes by indicating the number of boron atoms with a Latin or Greek prefix and the number of hydrogens with a numerical suffix. Examples are given in Table 4.

Ternary compounds. Many ternary compounds (compounds of three elements) can be considered as oxygen acids or their salts. The names of most common oxygen acids are nomenclature systems. Familiarity, however, requires retention of these names. Systematic names can be assigned on the basis of coordination-compound nomenclature (see below *Coordination compounds*); these, however, often are so cumbersome that the common names are preferred. A list of common names of well-known

Table 4: Names and Formulas of Some Boranes

formula	name
B_2H_6	diborane-6
B_5H_9	pentaborane-9
B_5H_{11}	pentaborane-11
$B_{18}H_{22}$	octadecaborane-22

ternary acids—and the related anions (negative ions)—is included in Table 5.

Salts—ternary compounds involving other elements in place of the hydrogen of the parent acids—are named according to the rules for binary compounds, with the anion from the acid playing the role of the more electronegative constituent. If one or more atoms of hydrogen remains, hydrogen is named as a second electropositive constituent. Examples of such salts are given in Table 6.

The occurrence of water, or similar molecules, in the composition of a material is indicated by adding to the end of the name the word hydrate for water (or, for example, ammoniate for ammonia) with a Latin or Greek prefix to indicate the number of molecules present. Examples of names of materials of this kind are given in Table 7.

Table 5: Common Names of Acids and Their Anions

formula	name of acid	name of anion	formula of anion
H_3BO_3	orthoboric acid*	borate	BO_3^{3-}
H_2CO_3	carbonic acid	carbonate	CO_3^{2-}
HNO_3	nitric acid	nitrate	NO_3^-
HNO_2	nitrous acid	nitrite	NO_2^-
H_3PO_4	orthophosphoric acid*	phosphate	PO_4^{3-}
$H_4P_2O_7$	pyrophosphoric acid	pyrophosphate	$P_2O_7^{4-}$
$H_5P_3O_{10}$	triphosphoric acid	triphosphate	$P_3O_{10}^{5-}$
$(HPO_3)_n$	metaphosphoric acid	metaphosphate	$(PO_3^-)_n$
$(HPO_3)_3$	trimetaphosphoric acid	trimetaphosphate	$(PO_3^-)_3$
H_3PO_3	phosphorous acid	phosphite	HPO_3^{2-}
H_3PO_2	hypophosphorous acid	hypophosphite	$H_2PO_2^-$
H_2SO_5	peroxosulfuric acid	peroxysulfate	SO_5^{2-}
$H_2S_2O_6$	dithionic acid	dithionate	$S_2O_6^{2-}$
H_2SO_4	sulfuric acid	sulfate	SO_4^{2-}
H_2SO_3	sulfurous acid	sulfite	SO_3^{2-}
$H_2S_2O_3$	thiosulfuric acid	thiosulfate	$S_2O_3^{2-}$
$HClO_4$	perchloric acid	perchlorate	ClO_4^-
$HClO_3$	chloric acid	chlorate	ClO_3^-
$HClO_2$	chlorous acid	chlorite	ClO_2^-
$HClO$	hypochlorous acid	hypochlorite	ClO^-
$HMnO_4$	permanganic acid	permanganate	MnO_4^-

*Prefix often omitted.

Table 6: Names and Formulas of Salts Based on Ternary Acids

formula	name
KNO_3	potassium nitrate
$NiSO_3$	nickel(II) sulfite
$KHSO_4$	potassium hydrogen sulfate
$Na_2H_2P_2O_7$	disodium dihydrogen pyrophosphate

Table 7: Names of Hydrated and Ammoniated Compounds

formula	name
$Na_2SO_4 \cdot 10H_2O$	sodium sulfate decahydrate
$NiSO_4 \cdot 7H_2O$	nickel(II) sulfate heptahydrate
$PtCl_4 \cdot 2NH_3$	platinum(IV) chloride diammoniate

Coordination compounds. A coordination compound is composed of one or more complex structural units, each of which has a central atom bound directly to a surrounding set of groups called ligands. Nomenclature of coordination compounds requires a knowledge of these structural relationships; this topic is covered below in the section *Coordination compounds.*

STRUCTURAL CLASSIFICATION OF INORGANIC COMPOUNDS

Salts. The compounds formed by the chemical union of strong metals (elements at the left-hand side of the periodic table, of which the most metallic are the alkali metals, Group Ia) with strong nonmetals (elements at the right-hand side of the table, of which the most nonmetallic are the halogens, Group VIIa) provide the simplest examples of salts. These include sodium chloride (common table salt) and all the other compounds of alkali metals and halogens, called halides, such as lithium fluoride, potassium bromide, rubidium chloride, and cesium iodide, as well as the halides of many other elements—for example, magnesium fluoride, calcium chloride, and lanthanum fluoride. These halides are most clearly electrovalent in nature. The oxides of these elements (binary compounds with oxygen), the sulfides (binary compounds with sulfur), and the nitrides (binary compounds with nitrogen) are also electrovalent.

Most other binary compounds (compounds made of two elements) are either clearly composed of molecules (rather than ions) or else the character of their bonding is open to question. Commonly such materials are partially covalent in nature.

From F. Cotton and G. Wilkinson, *Advanced Inorganic Chemistry* (© 1966); John Wiley & Sons, Inc.

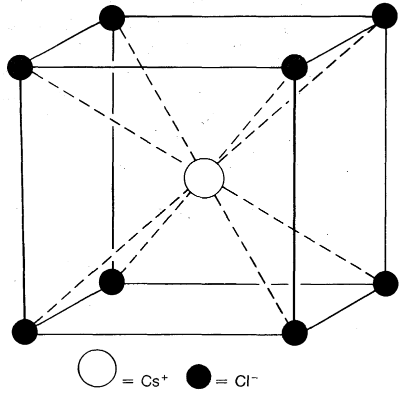


Figure 4: Cesium chloride structure.

Ionic crystal lattices

Electrovalent substances form solid crystals that are described in terms of ionic lattices. In these lattices, the positive ions are surrounded by negative ions, and the negative ions, in turn, are surrounded by positive ions. In any particular ionic lattice, a definite number of ions of one kind is arrayed about those of the second kind, and vice versa. This number is called the coordination number. In the sodium chloride lattice, the coordination number of both the cation (positive ion), the sodium ion, and the anion (negative ion), the chloride ion, is 6. The halide salts of most of the alkali metals crystallize in this same lattice form. The main exceptions are the cesium salts (*e.g.,* cesium chloride), which have a lattice in which the coordination number of both ions is 8 (see Figure 4).

The anions and cations have the same coordination numbers only in structures in which the same numbers of the two kinds of ions are present. In the lattice of calcium fluoride (CaF_2), the coordination number of the calcium cation is 8, whereas that of the fluoride anion is 4 (see Figure 5).

When salts melt, the geometric constraint of the crystal lattice is overcome, with the result that the individual ions become relatively free to move. The fact that the ultimate particles of the molten salt are charged makes it possible to pass an electric current through such melts. When such conduction occurs, the ions migrate through the liquid toward the electrode of opposite charge, thereby constituting an electric current. In water solutions, the covalent water molecules separate the ions so that they can move independently. Consequently, such solutions are also electric conductors.

Oxides, anhydrides, acids, and bases. The physical properties and chemical behaviour of the oxides of the elements provide a scheme for systematizing much of the chemistry of inorganic compounds. The gases of low atomic number—helium, neon, and argon—do not form stable oxides. The oxides of the early elements in any period of the table (such as sodium and calcium) are typically electrovalent, or saltlike, as evidenced by their high

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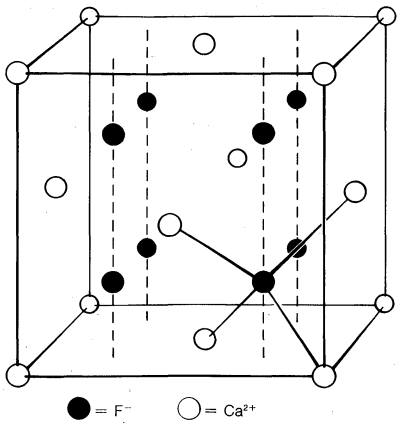


Figure 5: Calcium fluoride (fluorite) structure.

melting and boiling points. In contrast, the oxides of the last members of a given period in the table are characteristically discrete molecular structures held together by covalent bonds and with only weak forces between the molecules, as evidenced by their low melting points and high volatilities (for example, phosphoric anhydride and sulfur trioxide). The oxides of the elements in the centre of a period are intermediate in bond type and physical properties.

The contrast in physical properties of the oxides of the elements extends also to their chemical behaviour. Oxides of the alkali and alkaline-earth elements (and of a few other metals, the ions of which have low valences—that is, few charges) are basic in character and react with water to form ionic hydroxides. A good example of this behaviour is the action of sodium oxide on water to give sodium hydroxide:

$$\underset{\text{sodium oxide}}{Na_2O} + \underset{\text{water}}{H_2O} \longrightarrow \underset{\text{sodium hydroxide}}{2NaOH}.$$

asic
xides

Sodium hydroxide is composed of sodium (Na^+) and hydroxide (OH^-) ions. Basic (the chemical opposite of acidic) character is associated with the presence of hydroxide ions in aqueous solutions. Bases neutralize acids, change the colours of certain dyes (red litmus turns blue in bases), and absorb gaseous carbon dioxide to form carbonates. The reaction of sodium hydroxide with carbon dioxide to give sodium carbonate is shown below:

$$\underset{\text{sodium hydroxide}}{2NaOH} + \underset{\text{carbon dioxide}}{CO_2} \longrightarrow \underset{\text{sodium carbonate}}{Na_2CO_3} + \underset{\text{water}}{H_2O}.$$

Solutions of bases are slick to the touch because of the destructive action of bases on skin tissue; and they taste bitter (although, of course, tasting is not advisable).

cidic
xides

The oxides of the last members of a period, especially in high oxidation states (that is, high valence), are acidic in character. Aqueous solutions of acids contain hydronium ions (H_3O^+). They neutralize bases, turn blue litmus red, liberate carbon dioxide from carbonates, and react with active metals, forming salts and evolving hydrogen. They also have a sour taste. The reaction of a typical strong acid, sulfuric acid, with metallic zinc to give a salt, zinc sulfate, and hydrogen is shown below:

$$\underset{\text{zinc}}{Zn} + \underset{\text{sulfuric acid}}{H_2SO_4} \longrightarrow \underset{\text{zinc sulfate}}{ZnSO_4} + \underset{\text{hydrogen}}{H_2}.$$

Standard definitions of acids and bases recognize the typical behaviour mentioned above and characterize the neutralization reaction as the production of a salt and water. In fact, sodium hydroxide exists in solution as sodium and hydroxide ions, and nitric acid exists as hydronium and nitrate ions; the neutralization process involves the reaction of hydronium ions and hydroxide ions to form water molecules.

In addition to the oxides mentioned above, binary hydrides of strong nonmetals are also acids. In a solution of hydrogen chloride in water, for example, almost all of the hydrogen chloride has reacted with water to form the hydronium and chloride ions:

$$\underset{\text{water}}{H_2O} + \underset{\text{hydrogen chloride}}{HCl} \longrightarrow \underset{\text{hydronium ion}}{H_3O^+} + \underset{\text{chloride ion}}{Cl^-}.$$

Hydrochloric acid is a strong acid because of its almost complete dissociation to the ionic form. In solutions of acetic acid (vinegar), however, most of the acid is present as discrete molecules of acetic acid rather than as ions, and this substance is, therefore, a weak acid. Some acids have more than one replaceable—or acidic—proton (hydrogen ion): phosphoric acid (H_3PO_4), for example, has three, and sulfuric acid (H_2SO_4) has two.

Acid–base
theories

According to the more general Brønsted acid–base theory (which is named after its first proponent, the Danish chemist Johannes Nicolaus Brønsted), an acid is defined as a proton donor, whereas a base is a proton acceptor. Neutralization is then the transfer of a proton from an acid to a base. In the reaction of gaseous ammonia (NH_3)—a base—with gaseous hydrogen chloride (HCL)—an acid—to form electrovalent ammonium chloride (NH_4Cl), neutralization has produced a new acid, NH_4^+, and a new base, Cl^-:

$$\underset{\text{ammonia}}{NH_3} + \underset{\text{hydrogen chloride}}{HCl} \longrightarrow \underset{\text{ammonium ion}}{NH_4^+} + \underset{\text{chloride ion}}{Cl^-}.$$

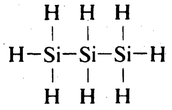
The new acid and base are both weaker than were the original reactants. Because of neutralization, the strongest acid that can exist in water is the hydronium ion (H_3O^+); similarly, the strongest base that can exist in water is the hydroxide ion (OH^-).

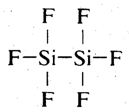
Another useful theory of acids and bases is the Lewis theory (named for the U.S. chemist Gilbert Newton Lewis), which specifies that a base is an electron-pair donor and an acid is an electron-pair acceptor. The "donated" electron pair is shared by the donor and the acceptor. According to this definition, water ($H_2\ddot{O}:$), hydroxide ion ($H\underset{\cdot\cdot}{\ddot{O}}:^-$), ammonia ($:NH_3$), and chloride ion ($:\underset{\cdot\cdot}{\ddot{Cl}}:^-$) are all Lewis bases. The most obvious Lewis acids (other than the proton H^+) are ions and atoms that have fewer electrons than required by the rule of eight. Boron trifluoride (BF_3), for example, has only six valence electrons and is capable of combining with a fluoride ion ($:\underset{\cdot\cdot}{\ddot{F}}:^-$) to form the tetrafluoroborate ion (BF_4^-). The boron atom in this ion obeys the rule of eight.

The oxides that form acids in aqueous solutions traditionally have been called acid anhydrides. The reaction of such an anhydride with water produces an acid, as in the following example with sulfur trioxide, SO_3:

$$\underset{\text{sulfur trioxide}}{SO_3} + \underset{\text{water}}{H_2O} \longrightarrow \underset{\text{sulfuric acid}}{H_2SO_4}.$$

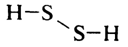
Catenates. The ideal covalent bond is a nonpolar one between two atoms of the same element; in this case, the electron pair is equally attracted to both atoms. The phenomenon of like atoms joining together by covalent bonds is called catenation (Latin *catena,* "chain"), and it reaches a maximum with carbon, giving rise to a great many compounds of this type. The other elements of Group IVa in the table show increasingly diminished abilities to undergo catenation (silicon much less than carbon, germanium and tin somewhat less than silicon, and lead least of all). Silicon forms the series of compounds denoted by the general formulas Si_nH_{2n+2} and Si_nX_{2n+2+} (in which n is any integer and X = Cl, F). The structures below represent low-molecular-weight examples of the two series:



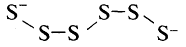


Catenation in Groups Va and VIa

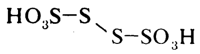
The elements of Group Va of the table show relatively little tendency to catenate. Hydrazine ($NH_2{-}NH_2$) is the only hydride of nitrogen containing only nitrogen—nitrogen single bonds. Among Group VIa elements, sulfur is unique in its ability to self-link; it forms several series of linear covalent molecules of the types represented by the following general formulas: HS_nH, ClS_nCl, and $HO_3SS_nSO_3H$. Salts of polysulfide anions (S_n^{2-}) are also well characterized. Structures are given below for representative compounds in these various classes:



dihydrogen disulfide



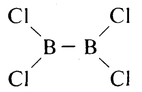
polysulfide anion



tetrathionic acid

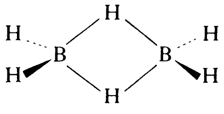
Some of the polysulfide ions are able to form chelate rings in which the sulfur atoms at each end of the chain are coordinated to one metal ion.

Boron is unique among Group IIIa elements in its ability to form catenates; the self-linkage of boron, however, is not restricted to typical boron–boron covalent binding. A few such compounds do exist, including diboron tetrachloride:



diboron tetrachloride

Although the simple hydride diborane (B_2H_6) was once thought to have a similar structure, it actually involves a distinctive three-centre, electron-deficient bond, in which one electron pair is shared between three atoms, two borons and one hydrogen atom:



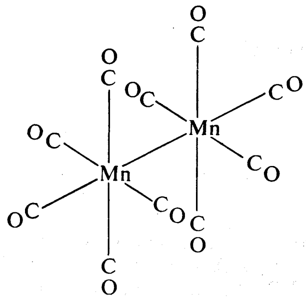
diborane

The ability of boron to form three-centre bonds in addition to normal covalent linkages leads to the formation of a number of boron hydrides with complicated structures. Although the simplest boron hydrides (for example, diborane) are extremely reactive toward oxygen and all solvents containing replaceable protons (such as water and alcohols), certain of the anions with cagelike structures (such as $B_{10}H_{10}^{2-}$ and $B_{12}H_{12}^{2-}$) are remarkably unreactive.

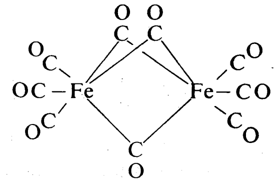
The carboranes are compounds similar to these anions except that two carbon atoms have replaced two of the boron atoms. Several of the carboranes with basket-like structures form pi complexes with metal ions that are very similar to the organometallic compounds with pi bonds described above.

Cluster compounds

Compounds in which the self-linking of elements has produced compact structures are referred to as cluster compounds. Catenation of metallic elements can yield both simple linear units and clusters. Though a number of representative metals apparently can produce metal–metal bonds under the right conditions, only mercury forms such species in aqueous mediums. The diatomic mercurous ion (Hg_2^{2+}) is unique in this regard. Given the right ligands (attached groups), however, most transition elements (those in the centre of the periodic table) can form metal–metal bonds. The carbonyls of iron and manganese provide examples of metal–metal bonds that are, respectively, accompanied by, and not accompanied by, bridging groups, as shown below:



decacarbonyldimanganese

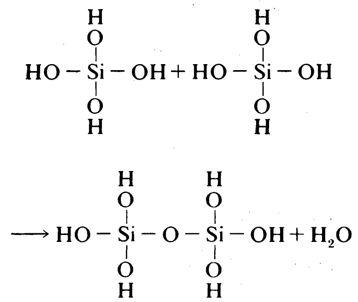


tri-μ-carbonylbis(tricarbonyliron)

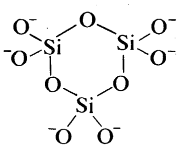
In the case of the rhenium halides with the formula $Re_2X_8^{2-}$, molecular orbital theory has been used to explain the unusually short rhenium–rhenium-bond distance in terms of a quadruple bond between the two rhenium atoms. Clusters occur with mutual bonding of three metal atoms [$Os_3(CO)_{12}$, Nb_3Cl_8], four metal atoms [$Co_4(CO)_{12}$], and six metal atoms [$Rh_6(CO)_{16}$ and $Mo_6X_8^{4+}$].

Polymeric silicates

Inorganic polymers. A polymer is a substance with a molecular structure containing repeating units; commonly, molecules with varying numbers of the repeating units are present, with the result that most polymer samples exist as a distribution of molecules having different molecular weights. Whereas the repeating methylene unit (CH_2) is the simplest and most familiar among organic polymers, the unit $-O{-}SiY_2-$ (in which $Y = O^-$ or OH) is the parent of many inorganic structures, especially certain minerals. Orthosilicic acid, $Si(OH)_4$, polymerizes by the elimination of water to form a vast array of structures, known as polymeric silicates, as shown by the following equation for the formation of a double molecule, or dimer; repetition of the process leads to progressively longer molecules.



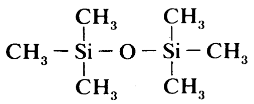
These polymers are best known in certain minerals (see the *Macropædia* article MINERALS AND ROCKS: *Silicates*). Long chains of $-SiO_3-$ units, for example, are found in the minerals enstatite, which conforms to the general formula $MgSiO_3$, and in spodumene, $LiAl(SiO_3)_2$. Other minerals, called amphiboles, and including such substances as tremolite, $Ca_2Mg_5(Si_4O_{11})_2(OH)_2$, contain double-stranded long chains in which the repeated unit is $S_4O_{11}^{6-}$. These substances are fibrous in character (asbestos being a good example), and the chains are held together by the cations. Mica and talc are two-dimensional sheet polymers of silicon and oxygen, whereas feldspar and zeolites are rigid, three-dimensional ploymers of similar type. Silicates also form such cyclic materials as $Si_3O_9^{6-}$, the structure of which is shown below:



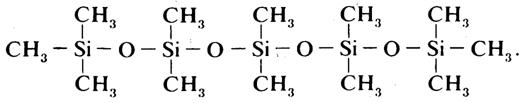
cyclic silicate anion

The most useful of man-made inorganic polymers are the siloxanes (also known as silicones), which use the basic structural unit of the silicate minerals, $-OSiY_2-$, with—

however—the Y's being methyl groups. Simple examples of these structures are:



$$CH_3-\underset{CH_3}{\overset{CH_3}{Si}}-O-\underset{CH_3}{\overset{CH_3}{Si}}-CH_3$$



$$CH_3-\underset{CH_3}{\overset{CH_3}{Si}}-O-\underset{CH_3}{\overset{CH_3}{Si}}-O-\underset{CH_3}{\overset{CH_3}{Si}}-O-\underset{CH_3}{\overset{CH_3}{Si}}-O-\underset{CH_3}{\overset{CH_3}{Si}}-CH_3.$$

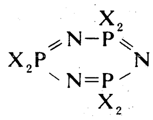
Siloxanes are chemically unreactive and thermally stable.

Phosphoric acid (H_3PO_4) and sulfuric acid (H_2SO_4) resemble orthosilicic acid (H_4SiO_4) in undergoing polymerization reactions, but the polymers designated by the formulas $H_2S_2O_7$ and $H_2S_3O_{10}$ are extremely sensitive to moisture. Sulfur trioxide, SO_3, exists in trimeric cyclic and linear polymeric forms.

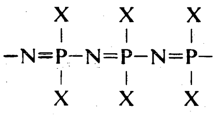
Polyphosphates and polysulfates

The polyphosphates are somewhat more stable than polysulfates, and long chain structures similar to the single-stranded silicates are known. The lower molecular weight anions, $(O_3P-O-PO_3)^{4-}$ and $(O_3P-O-PO_2O-PO_3)^{5-}$, are the most important. Cyclic phosphates are also known. Tungsten and molybdenum form a remarkable series of polyacids in which cluster-type polymerization occurs through oxygen bridges.

Phosphorus forms a series of compounds of the formula $(PNX_2)_n$, which can exist in both cyclic and linear polymeric forms, as shown below:



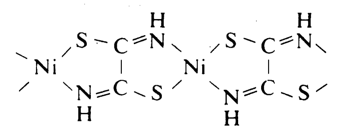
cyclic form



$$-N=\underset{X}{\overset{X}{P}}-N=\underset{X}{\overset{X}{P}}-N=\underset{X}{\overset{X}{P}}-$$

linear polymeric form

Polymers held together by coordination have also been prepared. A substance that found early use in producing coated polarizing lenses had the structures shown:



Unusual valence states. Oxygen forms metal salts containing the peroxide ion O_2^{2-}, the superoxide ion O_2^{-}, or the ozonide ion O_3^{-}. The last two are powerful oxidizing agents, whereas the peroxide ion can act either as an oxidizing or as a reducing agent.

Oxygen fluorides

Oxygen is assigned positive oxidation states only when it is combined with fluorine, the most electronegative of all elements. The compounds formulated as OF_2, O_2F_2, O_3F_2, and O_4F_2 are all strong oxidizing agents, and all are likely to react violently with most other substances. Trioxygen difluoride, O_3F_2, for example, is even more reactive than elemental fluorine. The great oxidizing strengths of these materials has led to interest in them as potential oxidizers for rocket fuels.

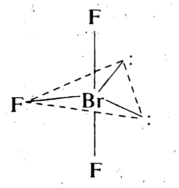
The binary compounds of hydrogen with elements at the right of the periodic table (Group IVa to VIIa) have already been discussed. The hydrides of the strong metallic elements are typical salts containing the hydride ion, H^-. The alkali-metal hydrides (*e.g.*, lithium and sodium hydride) have the same crystal structure as sodium chloride. The hydride ion also forms many complexes, especially with the electron-deficient hydrides of the Group IIIa elements. The formation of lithium borohydride is shown below:

$$\underset{\text{lithium hydride}}{2LiH} + \underset{\text{diborane}}{B_2H_6} \longrightarrow \underset{\text{lithium borohydride}}{2LiBH_4} .$$

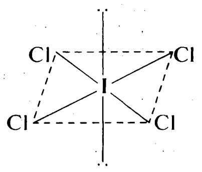
The complex hydride lithium aluminum hydride is an especially important reducing agent (that is, one that provides electrons) in chemical syntheses.

Interhalogen compounds

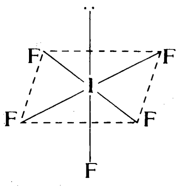
The halogens react among themselves to form a remarkable series of molecular interhalogen compounds (such as iodine chloride, bromine trifluoride, and iodine pentafluoride) and interhalogen ions (such as those formulated as I_3^-, ICl_4^-, BrF_4^-). The structures of these substances have been of theoretical importance because unshared pairs of electrons on the central atoms play a part in determining the geometries of the molecules, as shown in the following diagrams:



bromine trifluoride



iodine tetrachloride



iodine pentafluoride

The fluorides of sulfur contrast strongly with those of oxygen. Sulfur tetrafluoride is a moderate fluorinating agent, and the hexafluoride can be used effectively as a dielectric material (one that does not conduct electricity) because of its great thermal stability and chemical inertness.

Boron and nitrogen are on opposite sides of carbon in the periodic table, and they occur in one-to-one ratios in a series of materials that resemble elemental forms of carbon (diamond and graphite) and certain hydrocarbons (see Figure 6).

The compounds of sulfur and phosphorus, tetraphosphorus trisulfide (P_4S_3), tetraphosphorus pentasulfide (P_4S_5), tetraphosphorus heptasulfide (P_4S_7), and tetraphosphorus decasulfide (P_4S_{10}), appear to have strange formulas; they can all be understood, however, on the basis of a single structural principle. White phosphorus is composed of discrete P_4 tetrahedra, and the various sulfides are derived from these by the addition of sulfur atoms either along the edges or at the apexes of the structure (as shown below), all such locations being filled in tetraphosphorus decasulfide:

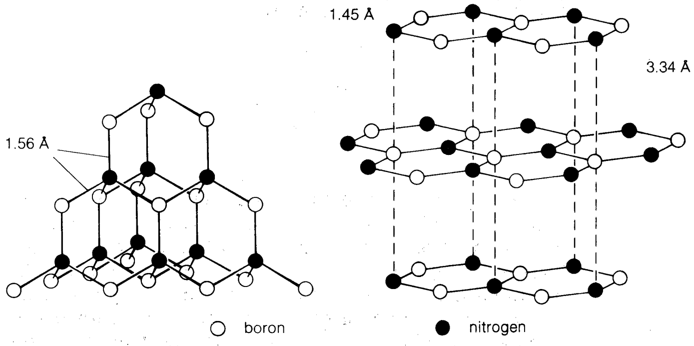


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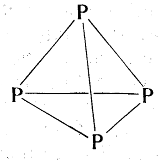
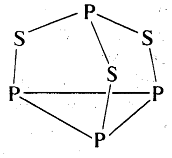
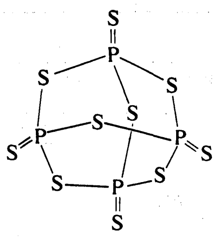
Figure 6: The structure of boron nitride, BN, illustrating its macromolecular nature and similarity to diamond and graphite. (Left) The arrangement of boron and nitrogen atoms in the diamond form of BN. (Right) The arrangement of boron and nitrogen atoms in the graphite form of BN.



white phosphorus



tetraphosphorus trisulfide



tetraphosphorus decasulfide

The anhydride of phosphoric acid, P_4O_{10}, has the same type of structure as tetraphosphorus decasulfide (with oxygen atoms replacing sulfur atoms). (D.H.B./Ed.)

WATER

Water is an extraordinary substance, anomalous in nearly all of its physical–chemical properties and easily the most complex of all the familiar substances that are single chemical compounds.

Ancient man, who clearly sensed the importance of water, made it a central feature of a great many mythological and religious cosmogonies, including that of Genesis. For the Greeks, the Titans Oceanus and Tethys were the parents of creation; and Western scientific tradition began in Greece when the Ionian philosopher-scientist Thales of Miletus, who flourished in the 6th and 7th centuries BC, substituted natural forces for gods as the causes of natural phenomena but retained water as the central principle, or "element," of his cosmology. His successors added other primal "elements"—fire, earth, and air—and the influence of Aristotle, who taught the doctrine of the four elements, adding a fifth element of his own (quintessence), dominated Islāmic and Christian thought. For many centuries the elements as qualities formed the basis of alchemical beliefs. Even Newton, an avid alchemist, set forth in *De Natura Acidorum* the view that all substances can be reduced to water. Eventually, hydrogen, one of the components of water, was considered to be the primal element, largely because it was found to be the lightest of all elements. The hypothesis that all the chemical elements were merely different aggregates of hydrogen was not abandoned until the 19th century, when precise determination of the atomic weights of the other chemical elements showed that they were not simple multiples of that of hydrogen.

Water as one of the alchemical elements

Structure of water. The formula for water, consisting of two atoms of hydrogen (H) and one atom of oxygen (O), is, thus, H_2O. The three atoms (*i.e.,* these three nuclei and their attendant electrons) are covalently bonded and form a specific structure, with the oxygen atom between the two hydrogen atoms.

The three atoms do not lie in a straight line, however, but, rather, with the hydrogen atoms bent toward each other. The three-dimensional structure of the water molecule, therefore, is unsymmetrical and can be shown with four uneven arms contained by a not quite perfect cube (Figure 7A). The angle formed by the two hydrogens and the central oxygen atom is about 105°, whether in the gaseous, liquid, or solid states, and the distance between the oxygen and each hydrogen nucleus—*i.e.,* the oxygen–hydrogen bond distance—is 0.96 angstrom (Å; one angstrom equals 10^{-8} centimetre). Opposite the hydrogen atoms and directed to the opposite corners of the containing cube are two electronic clouds of negative electrification, which are the keys to water's peculiar behaviour: they attract the hydrogen nucleus of an adjacent water molecule to form what is called a hydrogen bond. These bonds are not especially strong, but, since they orient the water molecules into a specific configuration, they significantly affect the properties of water. Water is not the only substance capable of hydrogen bonding. Ammonia, hydrogen fluoride, and alcohols also form hydrogen bonds; *i.e.,* hydrogen nuclei in these compounds are enough exposed to be attracted by the concentration of negative charges on adjacent molecules. Such molecules, so structured that one part has an excess of positive and the other an excess of negative charges, act somewhat like minute magnets and are called polar molecules. Polarity can also sometimes be induced in a molecule by the environment. The uniqueness of water appears to lie, however, in the ease with which its molecules form extended, three-dimensional, polymorphic hydrogen-bonded aggregates.

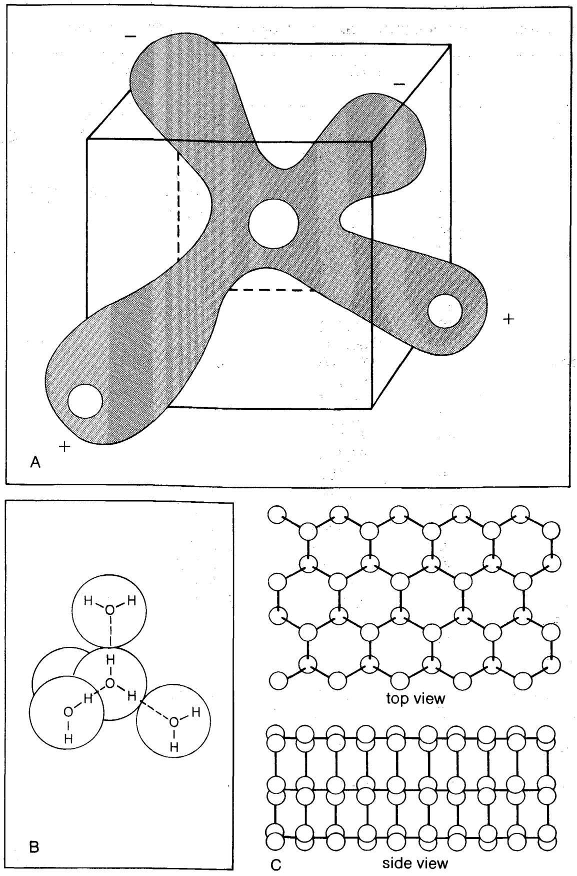


Figure 7: *The effect of the electronic structure of individual water molecules on the pattern of their organization in the crystalline state.*
(A) Electronic cloud of a single water molecule. (B) Tetrahedral arrangement of water molecules in the ice I lattice. (C) The structure of ice I.

Effects of hydrogen bonds. In the gaseous state, whether steam or water vapour, water molecules are largely independent of one another; and, apart from collisions, interactions between them are slight. Gaseous water, then, is largely monomeric; *i.e.,* it consists of single molecules, signified by a subscript, $(H_2O)_1$, with an occasional water dimer—*i.e.,* molecule formed by the union of two simple molecules, $(H_2O)_2$—and, perhaps, some even rarer trimers, consisting of molecules formed by the combination of three simple molecules, $(H_2O)_3$. In the solid state, at the other extreme, the water molecules interact with one another strongly enough to form an ordered structural lattice. A water molecule collects the four nearest of its neighbours and arranges them about itself in a tetrahedral configuration (Figure 7B). When the water freezes, this results in a lattice that can be schematically represented as sheets of puckered hexagonal rings (Figure 7C). Since the molecules in such a lattice are not crowded closely together, water is one of the few substances that is less dense in solid than in liquid state, and, for this reason, lakes and ponds develop a surface ice sheet in winter rather than freezing solid from bottom to top, as does a liquid that is less dense than its solid form (*i.e.,* normal liquids become more dense [they shrink] as the temperature is lowered, and the coldest layers will sink to the bottom, where they will be the first to freeze).

Ice

A variety of structures are formed when water freezes, and they are identified by roman numerals. The structure represented in Figure 7C is that of ice I, the familiar ice that forms and is stable at ordinary atmospheric pressure. Nine different forms of ice are known, each definable by

Table 8: Comparative Properties of the Forms of Water Compounds

	temp. (°C)	$(H_2O)_n$	$(D_2O)_n$	H_2S	normal pentane (C_5H_{12})	seawater	ice I	H_2O vapour
Molecular weight		18.015*n*	20.031*n*	34.08	72.15		18.015*n*	18.015
Density, g/ml	0	0.99987		2.12 (−42° C)			0.9168	
	5	0.99999	1.10562					
	15	0.99913	1.10587		0.626			
	25	0.99707	1.10445	3.6×10^{-3}		1.02412		
	40	0.99224	1.0996					
	100	0.95838						5.8×10^{-4}
Vapour pressure, mm Hg	0	4.580						
	5	6.543						
	15	12.788						
	25	23.756						
	40	55.324						
	100	760.00						
Dielectric constant	0	87.740	87.740				79 (−1° C)	
	5	85.763						
	15	81.945						
	25	78.303						
	40	73.181						
	100	55.720						
Viscosity, cP	0	1.787				1.877		
	5	1.518	1.981					
	15	1.139	1.434					
	25	0.8904	1.096			0.902		9.6×10^{-3} (20° C)
	40	0.6529	0.785					
	100	0.2818	0.785					
Melting point, °C		0.0	3.82	−85.5	−129.72	−1.91		
Boiling point, °C	14.203	100.0	101.42	−60.7	36.07	100.56		
Temp. at maximum density, °C		3.98			none	−3.52		
Surface tension, dyne/cm	25	71.97				72.8		
Index of refraction	25	1.3325	1.3388		1.357 (20° C)			0.380 (100° C, air)
Self-diffusion coefficient, cm³/sec	25	2.57×10^{-5}						
Heat of ionization, kJ/mole	25	55.71						
Heat of formation, kJ/mole	18	285.89	294.59				292.7 (0° C)	242.49 (100° C)
Heat of fusion, kJ/mole	0	6.010	6.345		0.116 (−130° C)			
Heat of vaporization, kJ/mole	100	40.651						
Apparent dipole moment, esu-cm	25	5.59×10^{-19}						
Sound velocity, m/sec	25	1,496.3						405 (100° C)
Isothermal compressibility, atm⁻¹	25	45.6×10^{-6}				42.2×10^{-6}	12×10^{-6} (0° C)	
Specific heat, J/g-°C	25	4.179					2.06 (0° C)	2.078 (100° C)
Thermal conductivity, watt/cm-°C	20	5.98×10^{-3}			1.35×10^{-3}	5.63×10^{-3}		2.44×10^{-4} (110° C)
Specific electrical conductivity, ohm⁻¹ cm⁻¹	25	$<10^{-7}$	$<10^{-8}$			0.0532		

its lattice structure. Changes in temperature and pressure transform one kind of ice into another. The changes are specific and have been charted in phase diagrams.

Liquid water

Although ice has many baffling aspects, the full implications of water's eccentricities are revealed by the liquid. Far from being a simple, or "normal," fluid with undissociated molecules moving randomly and more or less independent of one another, water exhibits a short-range ordering reminiscent of the long-range ordering in the ice lattice, because the hydrogen bond makes water molecules stick tenaciously to one another. This results in anomalously high values for viscosity, surface tension, and the temperatures and heats of melting and boiling, which involve separation of molecules (Table 8). By analogy with hydrogen sulfide, H_2S (a molecule similar to water chemically), water should be a gas rather than a liquid at room temperature.

The stickiness of water molecules was recognized in the 19th century, and early attempts to explain it were in terms of association to form small aggregates called dihydrol and trihydrol. Water is an associated liquid best represented by the formula $(H_2O)_n$, in which n represents the number of H_2O molecules included in the aggregate. At present, despite intensive research for many years, there is uncertainty about the exact value of n—it seems dependent on temperature and pressure and the type and concentration of solute present, and its value may depend on where the water sample has been taken; there is a difference between n at a surface and n deep in the bulk.

Many theories have been developed to explain the structure of pure liquid water; some can account for most of water's peculiar properties, but none can account for all. Certain theories envision the liquid as consisting of a partially broken-down ice I lattice; others define a continuum model, and still others project a mixture model. A popular example of the latter postulates the liquid as a mixture of clusters (or polymers) with an average value

Flickering clusters

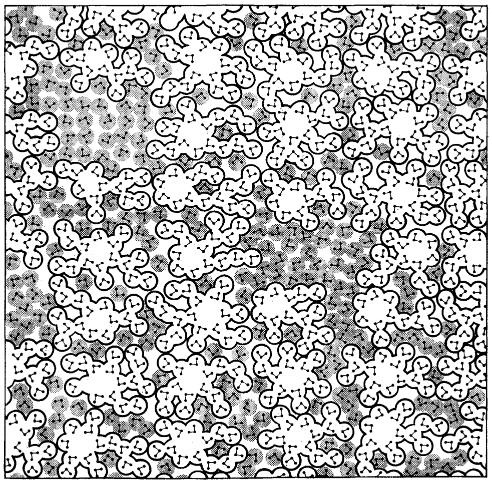


Figure 8: The "flickering cluster" model of liquid water showing a mixture of aggregates (outlined) and monomeric water (shaded).

for n of about 40 (at 20° C [68° F] and one atmosphere of pressure) and of monomeric ($n = 1$) water (Figure 8). The different groupings are in constant change, or flux, and are called flickering clusters. The average cluster size, n, decreases with increasing temperature or pressure, thus accounting for such properties as the maximum in the temperature dependence of the density and the minimum in the pressure dependence of the relative viscosity. The exact nature of the aggregates, if indeed there are clusters, is a subject of controversy.

The thermodynamics (*i.e.,* the study of heat and related energy changes) of the dissolution of nonpolar (molecules with no effective concentrations of charges—*i.e.,* no opposite poles) solutes in water indicate that they tend to order the water about themselves to form what has been described as an iceberg, the implication being that the structure is more or less analogous to that of ordinary ice. The implication, however, is unwarranted, and it is preferable to be less explicit and to describe this phenomenon as hydrophobic hydration. In some instances the crystalline hydrates (compounds or atoms that have combined with or added water molecules to themselves) of inert gases and other hydrophobic (water-repelling) solutes have actually been isolated and their structure determined. These gas hydrates (belonging to a class of molecules called clathrates) consist of a "guest" solute molecule enclosed in a polygonal cage of water molecules.

Hydration effects

Some ionic and polar solutes appear to make the water about them more structured, whereas others appear to be water-structure breakers. In order to account for this diverse behaviour, a process called Coulombic hydration is envisioned as producing the complex structure of what may be called a hydration envelope. First, there is an inner sphere of strongly bound molecules (called primary hydration); this forms part of a region of enhanced water structure, which in turn is surrounded by a zone of disrupted water structure (Figure 9). Depending on which zone is the more important, the solute will be a structure maker or breaker. So strongly held are the innermost water molecules that they may stick to the solute even when it is crystallized out of solution—the so-called water of hydration. It is possible to speak of hydration numbers, but there is little agreement in the values obtained by the different experimental techniques in use (Table 9). Recently, a Soviet school of physical chemists tried to circumvent this difficulty by thinking instead in terms of the time a water molecule will remain in a position near the solute species compared with the length of time a water molecule will remain in a position that is a great distance from the solute.

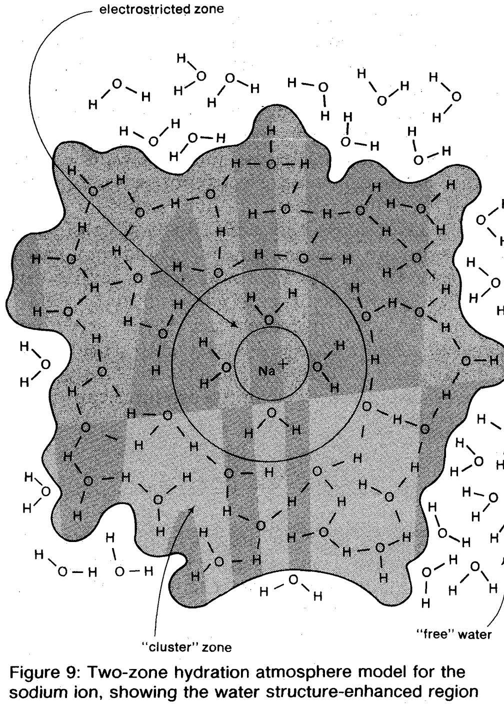


Figure 9: Two-zone hydration atmosphere model for the sodium ion, showing the water structure-enhanced region (the electrostricted zone plus the "cluster" zone) and the structure-broken region ("free" water).

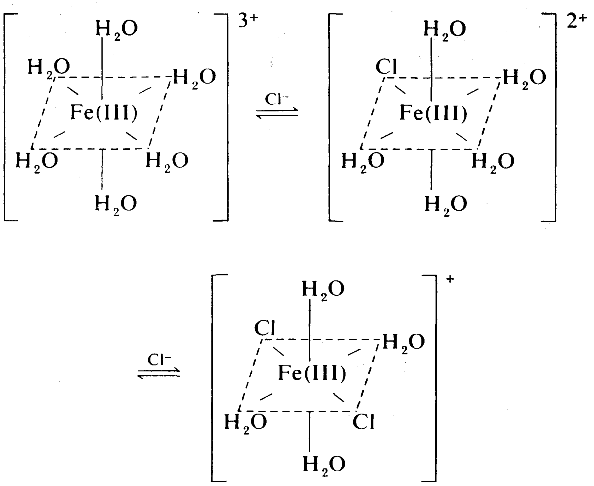
Table 9: Hydration Numbers of a Few Common Ions

ion	range of reported values	not unlikely value
Na^+	2–70	4
K^+	0.6–20	4
Mg^{2+}	6–36	12
Ca^{2+}	4–29	8
Br^-	2–5	2
SO_4^{2-}	1–12	12

Coordination in hydrates

Although it is an oversimplification to equate an ion's hydration number (number of water molecules associated with the ion) with its coordination number (the number of molecules, atoms, or ions, collectively called ligands, bonded to a central atom in a complex molecular structure), the hydration of an ion is closely related to its coordination chemistry. The water molecules can occupy the coordination positions around the central ion and be replaced with other ligands, either neutral molecules or ions.

The ion of iron(III), Fe^{3+}, has grouped around it six water molecules in a coordination compound that can be represented in a three-dimensional sketch:



Chlorine ions can replace the water, and, particularly if the ligands are organic groups, such replacement can result in drastic changes in chemical properties. If, for example, four of the six water molecules in the iron coordination compound are replaced by protoporphyrin, an organic compound, as is the case in heme (a portion of hemoglobin), the ability of the iron ion to act as a catalyst for the decomposition of hydrogen peroxide is increased 1,000-fold; if the remaining two water molecules are replaced by a protein compound to form the oxidizing enzyme catalase (which occurs in the blood and tissues), the catalytic activity is further increased 10,000,000-fold. The iron in heme, with two coordinated water molecules, is rapidly oxidized by molecular oxygen, but, if one of the two water molecules is replaced by protein, as in hemoglobin, the resulting compound can combine with oxygen reversibly and transport molecular oxygen from the lungs to distant cells without oxidation of the ferrous iron.

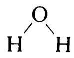
Generally speaking, the greater the charge density—*i.e.,* the ratio of charge to surface area—of an ion, the more heavily hydrated it will be, and as a rule negative ions (anions) are less heavily hydrated than positive ions (cations) because of the greater difficulty of crowding water molecules about the anions with the positive charges of the water protons oriented inward; *i.e.,* it is easier to orient the water with the protons outward and the negative end inward toward the positively charged cations.

Effects of interfaces

The structure of liquid water is also perturbed by the presence of an interface, but just how deep this layer of water (called vicinal) extends and what its structure might be are even less clear and more controversial than the structure of the bulk liquid. Vicinal water is suspected to have properties quite different from those of ordinary bulk water, a higher viscosity for example; this could be important since, if the layer is indeed thick, then most of the water in droplets in the atmosphere, most of the interstitial water in marine and other sediments, and most of the water in living cells and tissues would be vicinal water. The situation was complicated by the announcement of the discovery of a new form of water, orthowater (or polywater); however, the bulk of the present evidence points in the direction of vicinal water contaminated with various levels of impurities, rather than a new liquid form.

Physical properties of water. The physical–chemical properties of water are of great importance. Not only is the terrestrial environment largely aqueous, but also, because life originated in the oceans and evolved in Earth's hydrosphere, the basic processes of the life function are inextricably dependent upon water's peculiar properties.

Table 8 lists selected properties, at one atmosphere of pressure, of pure ordinary water, $(H_2O)_n$; pure heavy water, $(D_2O)_n$ (one isotope of hydrogen has a proton and a neutron in the nucleus, with resulting mass number 2, and is called deuterium; water composed of this heavier isotope of hydrogen is called heavy water); water's sister compound, H_2S; a "normal liquid, *n*-pentane; 35 parts per thousand salinity seawater; ice I; and water vapour. As has been indicated, the relatively high melting and boiling points, surface tension, and heats of fusion and vaporization are all the consequence of the disinclination of the hydrogen-bonded molecules to separate. The high viscosity also is a reflection of the extensively structured nature of liquid water; the very high dielectric constant (a dielectric substance acts as a nonconductor of electricity; the more insulating the substance, the higher is its dielectric constant), one property that makes water such a good solvent, is in part the result of the molecule being bent,



rather than linear, H—O—H, so that there is a strong dipole moment (measure of the tendency of a molecule with oppositely charged parts to be affected by an electrical, or magnetic field). Heavy water, D_2O, is slightly more structured than ordinary water; that is to say, the deuterium bond is slightly stronger than the hydrogen bond, so that many of its properties tend to have somewhat higher numerical values.

Chemical properties of water. Water is a highly reactive chemical substance, but, because it is recoverable from so many familiar chemical and physical processes, it was long thought to be elemental—*i.e.,* not a compound. Its compound nature was not demonstrated until quite late, and it was not directly synthesized from its elements until 1781. Because of early confusion between the atomic and molecular weights of gases, the formula of water was for some time erroneously thought to be HO rather than H_2O. Thermally, water is very stable; at 2,700° C (4,900° F) it is still only about 11 percent dissociated into hydrogen and oxygen molecules according to the equation (in which the number of each molecule involved is given before its formula and two arrows indicate that the reaction can go in either direction—*i.e.,* that it is reversible): $2H_2O \rightleftharpoons 2H_2 + O_2$.

Ionization of water

The self-ionization of water is small but extremely important. One molecule of water breaks up into a positive hydrogen ion, H^+, and a negative hydroxyl ion, OH^-, and again the reaction is reversible; *i.e.,* the ions recombine to form a water molecule: $H_2O \rightleftharpoons H^+ + OH^-$. According to the law of mass action, which pertains to reversible reaction in a system, when the two opposite reactions proceed at equal rates (when the production of ions equals the production of water)—*i.e.,* when the system is in a state called equilibrium—and the actual concentrations of the constituents (in this case H_2O, H^+, OH^-) are not changing, even though reactions among them are taking place, the ratio of the concentrations of the constituents equals a constant. If brackets are taken to indicate concentration, then the product of ion concentrations divided by the concentration of water equals a constant:

$$\frac{[H^+]\,[OH^-]}{[H_2O]} = \text{constant.}$$

Because so few molecules dissociate, so that the concentration of water molecules is always enormous, relative to its ions, it, $[H_2O]$ in the denominator, can also be considered a constant, K. Thus, $[H^+]\,[OH^-] = \text{constant} \times [H_2O] = K$. Measurements show that K equals 10^{-14}, which means that at room temperature pure water contains 55 moles (one mole is the weight in grams of the molecular weight—*e.g.,* 18 grams of water) of H_2O molecules per litre but only about 10^{-7} moles each of H^+ and OH^- ions. If the hydroxyl ion concentration $[OH^-]$ is changed, the hydrogen ion concentration $[H^+]$ always changes automatically to maintain the equation. It is thought that it is the hydrogen ion that gives an acid its acidic characteristics, while the properties of the hydroxyl ion are those of bases. Acidity is thus defined in terms of concentration of hydrogen ion and is expressed as the pH of a solution. The pH is 7 in a neutral solution or in pure water. This concentration of hydrogen ions is taken as the base for the chemical acidity scale. Systems with a pH less than 7 are said to be acidic; those with a pH greater than 7 are alkaline or basic. At one time, the proton, H^+, in solution was represented as associated with a water molecule in what was called hydronium ion, a complex resulting from the H^+ ion bonding with a water molecule and giving it a positive charge: H_3O^+. More recent evidence indicates that the proton in aqueous solution is more heavily hydrated and should be represented as bonded to four water molecules: $H_9O_4{}^+$ or $H(H_2O)^+{}_4$, which is to say that the primary hydration number of the proton is 4.

Acidity and pH

Acid–base reactions comprise a major class of chemical reactions. Another important class of chemical reactions is oxidation–reduction: a substance is oxidized when it loses electrons and is reduced when it gains electrons. Just as no acid can react as an acid unless a base is present, so no oxidation of one reactant can take place without the presence of a reactant that can be reduced. Hydrogen gas, H_2, is a reducing agent; *i.e.,* it is oxidized to hydrogen ion, H^+, and electron, e^-, according to the equation $H_2 \rightarrow 2H^+ + 2e^-$. The potential for any substance to be reduced or oxidized can be compared with the potential of any other, and therefore such substances can be arranged in a list according to increasing potential to be oxidized. The standard against which the potential of any substance can be measured is hydrogen gas, and thus water forms the basis of what is called the electromotive series, with the oxidation–reduction potential of hydrogen set at 0 (by international convention). Very active metals high in the series (such as sodium, Na) spontaneously (and sometimes violently) displace hydrogen gas (indicated by suffixed *g*) from water at room temperature to give sodium ion, Na^+, hydroxyl ion OH^-, and hydrogen gas, $H_2(g)$, according to the equation: $2Na + H_2O \rightarrow Na^+ + OH^- + H_2(g)$. Less reactive metals such as iron (Fe), react very slowly with water but more rapidly with steam to produce iron oxide and hydrogen gas:

$$3Fe + 4H_2O\ (\text{steam}) \overset{600^\circ\,C}{\rightleftharpoons} Fe_3O_4 + 4H_2(g).$$

Metals, such as silver, below hydrogen in the series, will not displace hydrogen from water. The importance of acid-base and oxidation–reduction reactions in the study of water lies in the fact that, with few exceptions, the acidity and oxidative ranges—that is to say, the proton and electron concentration ranges—in the natural environment are those set by water and its reactions. Life has evolved within these ranges, and values of pH and E_H (oxidation potential) outside them rapidly prove fatal to most

Oxidation and reduction reactions

living things. The tolerance of organisms to temperature may also be set, at least in part, by water properties.

Water as a solvent

Although exotic chemicals are being synthesized constantly, water remains one of the best solvents known, perhaps closest to being the "universal solvent" of which alchemists dreamed. All chemical substances have a finite solubility in water ranging from compounds like ethanol (ethyl alcohol), which is miscible with water in all proportions, to so-called insoluble substances like mercuric sulfide, saturated solutions of which contain less than 10^{-25} dissolved ions per litre. The high dielectric constant (a measure of nonconductivity) of water greatly facilitates the dissolution of ionizable species. Equally important is the strength with which water molecules solvate (react with and take into solution) ionic and other solutes. The coexistence of oppositely charged species, such as sodium ion (Na^+) and chlorine ion (Cl^-), in water without recombination is testimony to the strength and stability of the hydration atmospheres (the molecules of water surrounding the solute) of solutes in aqueous solution. In addition to hydration, represented by the equation

$$Na^+ + (n + 4)\,H_2O \leftrightharpoons Na\,(H_2O)_4\,(H_2O)_n{}^+,$$

there is also the hydrolysis of metallic species, a reaction in which the metal ion displaces the proton of water to form basic substances. The following equations illustrate the step by step hydrolysis of iron(III) ion to iron(III) hydroxide and hydrogen ion:

$$Fe^{3+} + H_2O \leftrightharpoons [FeOH]^{2+} + H^+$$

$$[FeOH]^{2} + H_2O \leftrightharpoons [Fe(OH)_2]^+ + H^+;$$

$$[Fe(OH)_2]^+ + H_2O \leftrightharpoons Fe(OH)_3 + H^+.$$

When these successive reactions are summed, the result can be stated as: $Fe^{3+} + 3H_2O \rightleftharpoons Fe(OH)_3 + 3H^+$. The hydroxides of iron are all bases.

Nonmetallic species, on the other hand, on being hydrolyzed, form acids, of which carbon dioxide is a common example. Ignoring the step by step process, the final result of the hydrolysis is hydrogen ion and the bicarbonate ion, HCO_3^-, according to the equation

$$CO_2 + H_2O \leftrightharpoons H^+ + HCO_3^-.$$

Some of these reactions have profound environmental significance. The hydrolysis of iron is intimately mixed with the chemistry of trace metals, including heavy metal pollutants such as mercury in natural waters, and forms the chemical basis for the production of the troublesome, highly acidic waters that often drain from old mining operations. As another example, if the rate of any of the steps for the dissolution of carbon dioxide gas in the oceans is slow enough, the accumulation of carbon dioxide in the atmosphere from the combustion of fossil fuels could result in a general warming up of the earth's surface in what is called the greenhouse effect. Theoretically, then, the polar ice cover could melt, and all the coastal cities disappear beneath the rising surface of the seas.

Water as a medium and a catalyst for chemical

Dissolution generally greatly enhances the chemical reactivity of substances; it is not surprising that, in part due to its good solvent properties, water is the medium and also a catalyst for countless chemical reactions. With the exception of petrochemical and combustion reactions, there are few chemical processes in our environment that do not occur in aqueous systems, including man's own respiratory and metabolic functions. Even the cooking of food takes place in water or involves dehydration and subsequent changes of proteinous material.

The rates of the fastest chemical processes in aqueous mediums are determined simply by the encounter frequency (frequency of collision) of the reactants. Consequently, the speed with which solute and solvent species can move through liquid water is a question of some interest. All the major transport processes in water—that is to say, processes that involve the movement of species within the liquid, such as electrical conduction, diffusion, viscous flow, and dielectric relaxation—require about the same energy, thereby implying that they all take place by the same mechanism, and the slow, rate-determining step of that mechanism is now believed to be the formation of a vacancy, or "hole," in the liquid water. Hole formation is then followed by the jump of a nearby solute or solvent species into the hole. The observed energy associated with this process is presumably that necessary to break hydrogen bonds and form the hole. These various movements take place in liquid water at great speeds, ranging from about 10^{-12} to 10^{-10} second. The average lifetime of the aforementioned flickering clusters falls in the same range.

(R.A.Ho.)

ORGANIC CHEMICAL COMPOUNDS

As previously observed, organic chemical compounds can be defined roughly as compounds of carbon. They are of great importance to mankind—indeed, to all forms of life on Earth—if only because most of the substances of which living things are composed fall into this category. The chief foodstuffs—namely, fats, proteins, and carbohydrates—are organic compounds, as are such vital materials as hemoglobin, chlorophyll, enzymes, hormones, and vitamins. Other materials that add to the comfort, health, or convenience of man are composed of organic compounds: clothing of cotton, wool, silk, and synthetic fibres; common fuels, such as wood, coal, petroleum, and natural gas; components of protective coatings, including varnishes, paints, lacquers, and enamels; antibiotics and synthetic drugs; natural and synthetic rubber; dyes; plastics; and pesticides.

Many of these synthetic organic compounds, as well as many waste compounds produced during their manufacture, are dumped eventually on the land and into the rivers, where they resist the chemical reactions that break down naturally produced organic structures. In addition, many noxious synthetic compounds, designed for the control of plants and insects, also are impervious to natural agents, and even low concentrations of these compounds threaten all forms of life. Thus, an increasing problem in organic chemistry is the ultimate disposal of synthetic products.

A recent trend in organic chemistry has emphasized greater knowledge of the mechanisms of reactions, a field dominated by the methods of physical chemistry. A trend in biology depends upon increasing use of knowledge concerning molecular structure. Thus, organic chemistry is now a field in which knowledge of physics and knowledge of biological processes meet.

Uniqueness of the carbon atom

The carbon atom is unique among elements in its tendency to form covalent bonds, not only with other elements but also with itself. Because of its position midway in the second horizontal row, or period, of the table, carbon is neither an electropositive nor an electronegative element; it therefore is more likely to share electrons than to gain or lose them. Moreover, for elements in its period, carbon has the maximum number of outer shell electrons (four) capable of forming covalent bonds. (Other elements, such as phosphorus and cobalt, are able to form five and six covalent bonds, respectively, with other elements, but they lack carbon's ability to bond indefinitely with itself.) The four bonds of the carbon atom are directed to the corners of a tetrahedron (a solid figure with four corners) and make angles of about 109.5° with each other. The result is that carbon atoms not only can combine with one another indefinitely to give compounds of extremely high molecular weight, but the molecules formed have a three-dimensional structure that leads potentially to an infinite variety of different molecules. It is the abundance of the covalent compounds of carbon, as well as their unique character, that justifies their characterization as a separate group, called organic compounds, whereas the compounds of all other elements are classed together as the inorganic compounds.

The elements other than carbon that usually are present in organic compounds, in the order of decreasing frequency, are hydrogen, oxygen, nitrogen, chlorine, and sulfur; in addition, many compounds are known that contain other elements.

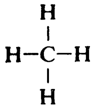
CLASSIFICATION AND NOMENCLATURE

Fortunately, despite the large number of organic compounds, a relatively simple systematic nomenclature has been adopted internationally that can be applied to the less complex—and hence the great majority of—organic compounds. This system was inaugurated by an international congress held in Geneva in 1892 and modified and extended at a meeting of the International Union of Chemistry held at Liège, Belgium, in 1930; since 1949, further development of the system has been continued under the auspices of the International Union of Pure and Applied Chemistry (IUPAC).

The IUPAC system

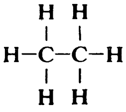
Hydrocarbons. Fundamentally, the IUPAC system is based on a nomenclature for the hydrocarbons (compounds containing only the two elements carbon and hydrogen).

The structures of organic compounds commonly are represented by structural formulas, which show not only the kinds and numbers of atoms present in the molecule but also the way in which the atoms are linked by the covalent bonds—information that is not given by simple molecular formulas. (With most inorganic compounds, the use of structural formulas is not necessary, because only a few atoms are involved, and only a single arrangement of the atoms is possible.) In the structural formulas of organic compounds, dashes are used to represent the covalent bonds. Atoms of the individual elements are represented by their chemical symbols as in molecular formulas. Examples of structural formulas of simple hydrocarbons are given below:



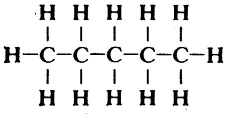
$$\begin{array}{c} H \\ | \\ H-C-H \\ | \\ H \end{array}$$

methane



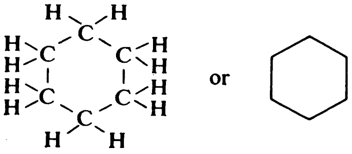
$$\begin{array}{c} H\ H \\ |\ \ | \\ H-C-C-H \\ |\ \ | \\ H\ H \end{array}$$

ethane

 or $CH_3CH_2CH_2CH_2CH_3$

$$\begin{array}{c} H\ H\ H\ H\ H \\ |\ \ |\ \ |\ \ |\ \ | \\ H-C-C-C-C-C-H \\ |\ \ |\ \ |\ \ |\ \ | \\ H\ H\ H\ H\ H \end{array}$$

pentane



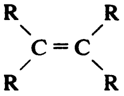
cyclohexane

Because so many hydrogen atoms appear in the various formulas when they are fully written out, it is customary to indicate simply the order of the carbon atoms with the number of hydrogens attached to each carbon indicated by a subscript as in molecular formulas, as shown in the alternate formula of pentane (above). Moreover, with cyclic compounds the rings are often indicated merely by polygons, each apex of which is considered to be a carbon atom bearing as many hydrogen atoms as needed for saturation.

In order to facilitate the naming of more complicated compounds, hydrocarbon groups that have one less hydrogen atom than the alkane are called alkyl groups, and their names have the same root as the corresponding alkane but end in the suffix -yl. Thus, CH_{3-} is known as a methyl group, and CH_3CH_{2-}, or simply C_2H_{5-}, is ethyl. The structure and nomenclature of the hydrocarbons are fully covered in the major section *Hydrocarbons* below.

Compounds with functional groups. Atomic groupings within molecules that are responsible for the varying chemical properties of organic compounds, such as double and triple bonds, are known as functional groups. The further classification and nomenclature of organic compounds is based on these groups.

If the symbol R is used to represent either hydrogen or an alkyl group, the general classes of organic compounds can be represented by attaching the proper number of R groups to the functional groups that distinguish the various classes. Alkenes, for example, can be represented by the general formula shown below:



(If one of the R groups in a class must be hydrogen, it is so indicated.) The functional group is indicated by either a suffix or a prefix added to the name of the corresponding alkane, as shown in Table 10. Before a vowel, the terminal *e* of the alkane is elided. For simpler compounds, common names generally are used.

Polyfunctional compounds

Polyfunctional compounds have more than one functional group. Hydroxycarboxylic acids, ketocarboxylic acids, and aminocarboxylic acids are common natural products. More complicated compounds with numerous mixed functions frequently are encountered.

Major categories of organic compounds—as defined by the functional groups present in their molecules—are treated in the following major sections of this article: *Alcohols, phenols, and ethers; Aldehydes and ketones; Carboxylic acids and their derivatives; Organic halogen compounds; Organic nitrogen compounds; Organic phosphorus compounds;* and *Organic sulfur compounds.* Other large groups of organic compounds—traditionally identified on the basis of other criteria—are dealt with in the sections *Alkaloids; Isoprenoids; Steroids; Dyestuffs and pigments; Heterocyclic compounds;* and *Polymers.*

PREPARATION AND PURIFICATION

Sources of organic compounds

Organic compounds are either derived from natural products or prepared by synthesis. The natural sources of organic compounds are plants and animals or fossilized materials, such as coal and petroleum. Synthesis is the use of chemical reactions to build larger molecules from smaller ones or to rearrange the atoms in a molecule to give new molecules.

The destructive distillation of coal to produce coke yields a liquid mixture called coal tar, which once was the chief source of aromatic compounds. Thermal decomposition of ethane, propane, and higher petroleum products is used to give alkenes of lower molecular weight; and the catalytic reforming of gasoline and kerosene fractions (which results in cyclization and the removal of hydrogen) now also yields aromatic hydrocarbons.

Fats, carbohydrates, and proteins not only are important as food substances but also yield fatty acids, sugars, and amino acids. Cellulose from cotton or wood can be converted to synthetic fibres, plastics, and explosives. Many important medicinals, such as the alkaloids and vitamins, as well as some dyes, flavours, and perfumes, are extracted from plants. Synthesis, however, accounts for most organic compounds, the initial starting points for which are chiefly the alkenes and the aromatic hydrocarbons. The chemical reactions of the various classes of organic compounds, many of which are of synthetic utility, are described below (see below *Chemical reactions*).

Separation methods

Most natural products and the products of synthetic reactions are mixtures that must be separated into the pure components. The procedures most commonly used are distillation, crystallization, and extraction with solvents. For the more difficult separations, selective adsorption and the various types of chromatography, a separation procedure based on differential adsorption of substances on various solid supporting materials, are used. Electrophoresis and ionophoresis, two processes that depend on the relative rates of migration in an electric field, can be used for charged particles.

It probably is fair to say that no ordinary compound is ever entirely free of foreign molecules as impurities. The

Table 10: Some Classifications of Organic Compounds According to Functional Group

class	function and name	general formula	suffix	prefix	examples
Alcohols	$-OH$, hydroxyl	$R-OH$	-ol	hydroxy-	C_2H_5OH, ethanol; C_6H_5OH, phenol; $HOCH_2CH_2OH$, 1,2-dihydroxyethane
Halogen compounds	$-X=$chlorine, bromine, or iodine	$R-X$		chloro-, bromo-, or iodo-	$CH_3CH_2CHClCH_3$, 2-chlorobutane (*sec*-butyl chloride)
Ethers	$RO-$, alkoxyl	$R-O-R$		alkoxy-	$C_2H_5OCH_3$, methoxyethane (ethyl methyl ether)
Carboxylic acids	$-COOH$, carboxyl	$R-COOH$	-oic acid		CH_3COOH, ethanoic acid (acetic acid)
Ketones	$>C=O$, carbonyl	$R-CO-R$	-one	oxo- (keto-)	$CH_3CH_2CH_2COCH_3$, 2-pentanone $CH_3COCH_2CH_2COOH$, 4-oxopentanoic acid
Aldehydes	$-\overset{\overset{H}{\vert}}{C}=O$, aldehyde	RCHO	-al	formyl-	$CH_3CH_2CH_2CHO$, butanal (*n*-butyraldehyde) $HCOCH_2CH_2COOH$, 3-formylpropanoic acid
Esters	$-\overset{\overset{O}{\Vert}}{C}-OR$, alkoxycarbonyl	$R-COOR'$	alkyl... -oate		$CH_3COOC_2H_5$, ethyl ethanoate (ethyl acetate)
Amines	—NH_2, amino	RNH_2	-amine	amino-	$CH_3\underset{\underset{NH_2}{\vert}}{C}HCH_3$, 2-aminopropane (isopropylamine)
Thiols (mercaptans)	$-SH$, sulfhydryl	RSH	-thiol	mercapto-	C_2H_5SH, ethanethiol (ethyl mercaptan) $HSCH_2CH_2COOH$, 3-mercaptopropanoic acid
Sulfonic acids	$-SO_3H$, sulfonic acid	RSO_3H	-sulfonic acid		$CH_3CH_2CH_2SO_3H$, propanesulfonic acid

degree of purity of any particular substance depends on the effort that is made to remove the last traces of impurities, and the effort expended generally depends on the use to be made of the compound. Compounds prepared for use as food additives or medicinal agents, for example, may need more extensive purification than those intended merely for industrial use. Once a compound is obtained as homomolecular (that is, in a form in which all the molecules present are identical) as is reasonably possible, it may be analyzed qualitatively for the elements present and then quantitatively for each element to determine the relative amounts of it in the molecule. The final step in establishing the molecular formula of the compound in question is the determination of the molecular weight. Today, in modern chemical laboratories, all of these analytical operations are carried out by specialists in microanalysis rather than by organic chemists.

Determination of structure

Once the molecular formula of a compound is established, the structure of the compound—that is, the order in which the atoms are attached to each other—must be elucidated. First, the kind and number of functional groups are determined. Then the molecule is broken down stepwise by known chemical reactions to smaller molecules of known structure. From the various bits of evidence, a structure is postulated in much the same way as a jigsaw puzzle is put together. This process has been greatly accelerated since about 1940, due to the development of physical procedures that use sophisticated electronic apparatus to secure information concerning molecular structure. Therefore, the determinations of ultraviolet and infrared spectra (light-absorption patterns), nuclear-magnetic-resonance spectra (patterns produced by magnetic effects on atomic nuclei), and mass spectra (separations of compounds—and fragments produced from them—on the basis of their relative mass) all play an important role in investigating the structures of relatively complex molecules. The use of radioactive isotopes of the various elements is fairly common, particularly in the study of organic reactions and in biochemical research on the way in which organic compounds are synthesized by living organisms.

The final proof of structure of a new organic compound is the synthesis of the supposed structure by known reactions from smaller molecules of known structure. Chemists have been able to arrive at the structure and configuration of substances with extraordinarily complex molecules, such as cholesterol, $C_{27}H_{46}O$, penicillin G, $C_{16}H_{18}N_2SO_4$, and chlorophyll *a*, $C_{55}H_{72}N_4O_5Mg$, by investigating their chemical and physical properties—and those of their degradation products. They also have been able to rebuild these molecules step by step from smaller molecules and in the end obtain compounds identical in all respects with the original product. This fact is the most convincing evidence of the soundness of the theories concerning the ways in which the various atoms are attached to one another in organic molecules.

PHYSICAL PROPERTIES

Since all compounds with a given functional group have approximately the same chemical properties, the differences among members of a given class of organic compound are largely differences in their physical properties. Boiling points, melting points, and solubility in various solvents are used most frequently to characterize a particular compound. Density, vapour pressure, and viscosity also often are measured. The most characteristic physical property of an organic compound, however, is its infrared absorption spectrum—that is, the particular pattern of wavelengths of light absorbed in the infrared region of the electromagnetic spectrum. If two samples have identical infrared spectra, they almost certainly are the same compound.

Factors that raise boiling points

Boiling point. In general, the boiling points of compounds with similar structures increase with increasing molecular weight. For simple, monofunctional, straight-chain compounds, the boiling point increases by about 20° C (36° F) for each additional carbon atom. Alkanes with four carbon atoms or fewer are gases at room temperature. For the higher molecular weight compounds, the boiling point rises from 36° C (97° F) for pentane, C_5H_{12}, to 287° C (548° F) for hexadecane, $C_{16}H_{34}$. Branching of the carbon chain lowers the boiling point. Thus, 2-methylbutane boils at 28° C (82° F) and 2,2-dimethylpropane at 9.5° C (49° F).

With the exception of double bonds, functional groups in a compound cause the boiling point to be higher. The chief factors that raise the boiling points of compounds above those of the corresponding alkanes are hydrogen bonding (attraction through secondary bonds involving hydrogen atoms attached to oxygen or nitrogen) and dipole–dipole interaction (attraction between polarized molecules). Both of these factors cause a degree of adhesion between individual molecules, hence making it more difficult to separate them. Table 11 lists the boiling points of some simple compounds of approximately the same molecular weight in the order of increasing boiling points.

Table 11: Boiling Points and Specific Gravities of Some Simple Organic Compounds

common name	formula	molecular weight	boiling point (°C)	specific gravity at 20° C
n-Butane	$CH_3CH_2CH_2CH_3$	58	−0.5	0.58
Methyl ethyl ether	$CH_3OCH_2CH_3$	60	10.8	0.73
Ethyl chloride	CH_3CH_2Cl	64	12.3	0.90
Ethyl mercaptan	CH_3CH_2SH	62	35.0	0.84
n-Propylamine	$CH_3CH_2CH_2NH_2$	59	49.0	0.72
Propionaldehyde	CH_3CH_2CHO	58	49.0	0.81
Acetone	CH_3COCH_3	58	56.0	0.79
n-Propyl alcohol	$CH_3CH_2CH_2OH$	60	98.0	0.80

Melting point. The melting points of solids increase with an increase of the forces holding the molecules in the crystal lattice. These forces depend on how readily the molecules fit together, a process that is influenced by the functional groups present and by the size and shape of the molecules. Usually, the more symmetrical the molecule, the higher the melting point of the compounds. Other generalizations have been made for certain types of compounds, but few predictions can be made concerning the melting points of most compounds.

Solubility. The solubility of one organic compound in another is the result of the intermingling of like and unlike molecules, whether they be gases, liquids, or solids. The extent to which compounds are mutually soluble depends on how nearly the attractive forces between like molecules match the attractive forces between unlike molecules. Thus, gases are all miscible with each other, because there is little in the way of attractive force even between the like molecules of a particular gaseous substance. The liquid alkanes are miscible in all proportions because the attractive forces between the like alkane molecules are much the same as between the unlike molecules. The liquid alkanes dissolve solid alkanes of lower molecular weights, but the solubility of the solid compounds decreases with increasing molecular weight, chiefly because the force with which the molecules are held in the crystal lattice increases with increasing molecular weight.

At the other extreme, water molecules are practically insoluble in liquid alkanes, and alkanes in water. In these cases, hydrogen bonding (secondary attraction by way of hydrogen atoms) of the water molecules to each other is many times stronger than the attractive forces between water molecules and alkane molecules, and there is little tendency to intermingle.

On the other hand, the lower alcohols, amines, and carboxylic acids are soluble in water for much the same reason that sodium chloride is soluble in water. Water molecules form hydrogen bonds not only with themselves but also with the hydroxyl group of alcohols and the amino group of amines and can hydrate the carboxylic-acid group and its ions, thus tending to equalize the attractive forces between the different molecules. For straight-chain alcohols with more than four carbon atoms and carboxylic acids with more than six carbon atoms, hydration is less effective in equalizing the attractive forces, and solubility of these compounds in water rapidly decreases. These considerations give rise to the old rule of thumb that "like dissolves like"; that is, the more closely organic compounds resemble one another in composition and structure, the greater is their mutual solubility.

Density and viscosity. The density of alkanes is lower than that of all other organic compounds, all alkanes being lighter than water. For liquids, a commonly cited datum is the specific gravity at 20° C (68° F); that is, the density of the substance at 20° C compared to that of water at 4° C (39° F); among the alkanes, the specific gravity varies from 0.63 for pentane (C_5H_{12}) to 0.77 for hexadecane ($C_{16}H_{34}$). The specific gravity increases with the introduction of atoms of elements that have a higher atomic weight than carbon (see Table 11) and with the increase of the ratio of the mass of such atoms to the mass of the rest of the molecule. Thus, ethyl chloride (C_2H_5Cl, $Cl/C_2H_5 = 1.2$) has a higher specific gravity, 0.90, than ethyl alcohol, 0.80 (C_2H_5OH, $OH/C_2H_5 = 0.6$); and 1,2-dichloroethane, at 1.26 (ClC_2H_4Cl, $Cl/CH_2 = 2.5$), is denser than water.

The viscosity of liquids depends on the ease with which molecules can slip past each other. Viscosity usually increases with increasing chain length of organic molecules because of the increased attraction between the molecules and the increased possibility of mutual entanglement.

CHEMICAL REACTIONS

Equilibria and rates of reaction

Most organic reactions are reversible to a greater or lesser extent; that is, the products of the reaction react with one another to regenerate the original reactants. In equilibrium reactions, the most important features are (1) the position of equilibrium, namely the ratio of products to reactants when chemical equilibrium is reached, and (2) the rate of reaction, or the speed at which the reaction takes place. The position of equilibrium is governed by the relative concentrations of reactants, the temperature, and the free energy change for the reaction —that is, the relative energy content of the reactants and products. The rate of reaction is influenced by the nature of the reactants, their concentrations, the temperature, the presence of catalysts or inhibitors, and the type of solvent used (if the reaction is carried out in solution). Acid–base reactions that involve readily ionized acids, for example, are extremely fast. When covalent bonds must be broken, however, the rate of the reaction at room temperature usually is very slow.

Since the rates of chemical processes approximately double for each 10° C (18° F) rise in temperature, organic reactions usually are carried out at elevated temperatures. In those instances in which high temperatures adversely affect the position of equilibrium, catalysts are sought that will make it possible for the reaction to proceed at reasonable rates and at as low a temperature as possible. Because both the forward and reverse reactions depend on the concentrations of the reactants, a desired reaction often can be made to go to completion by removing one of the products of the reaction; water, for example, may be removed by distillation or by the addition of a dehydrating agent. Since about 1940, chemists have become increasingly aware of the effect of solvents on the rates of reactions. Rate increases of more than 1,000,000-fold have been produced merely by changing the solvent.

Types of reactions

Because of the large number of different parent organic compounds and different functional groups, many types of organic reactions are possible. Some of the more important types are acid–base, addition, substitution, displacement, hydrolysis, pyrolysis, condensation, polymerization, and molecular-rearrangement reactions.

Acid–base reactions. Acids may be defined as compounds that can transfer a proton (a hydrogen atom that lacks an electron) to a base and a base as any entity with an unshared pair of electrons (and therefore capable of accepting a proton). In acid–base reactions a proton is transferred from an acid to a base, as shown in the following equation

$$H{:}A + {:}B \rightleftharpoons H{:}B^+ + {:}A^-,$$

in which HA represents any acid and B any base.

If HA and B are neutral molecules, the product is a positive ion and a negative ion and is known as a salt. A specific example is shown below, in which acetic acid is the acid involved, ethylamine is the base, and ethylammonium acetate is the resulting salt.

$$\underset{\text{acetic acid}}{CH_3COOH} + \underset{\text{ethylamine}}{{:}NH_2C_2H_5} \longrightarrow \underset{\text{ethylammonium acetate}}{CH_3COO^- + {}^+NH_3C_2H_5}$$

Varying degrees of acidity—that is, variations in the ease of removal of the proton—are possible, ranging from the extremely strong sulfonic acids to the virtually nonacidic alkanes; and varying degrees of basicity—or attraction for the proton—ranging from the extremely strongly basic alkide ion, R^-, to the extremely weakly basic sulfonate ion, RSO_2O^-, are also possible.

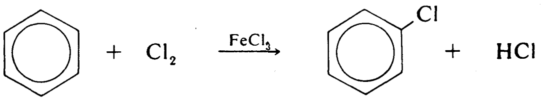
Degrees of acidity and basicity

Addition reactions. The addition of one molecule to another to give a single new molecule is an important class of reactions. Illustrative is the addition of bromine to ethylene.

$$H_2C=CH_2 + Br_2 \longrightarrow H_2CBrCH_2Br$$

ethylene bromine 1,2-dibromoethane

Substitution reactions. The replacement of hydrogen on carbon by another group generally is referred to as substitution. A typical substitution reaction is the synthesis of chlorobenzene from benzene and chlorine using ferric chloride as catalyst, as shown below:



benzene chlorine ferric chloride chlorobenzene hydrogen chloride

Displacement reactions. The direct replacement of one functional group by another generally is classed as a displacement reaction. A simple displacement of one halogen atom (chlorine) by another (iodine) in acetone as solvent is shown below:

$$CH_3CH_2CH_2Cl + Na^+I^- \longrightarrow CH_3CH_2CH_2I + Na^+Cl^-.$$

propyl chloride sodium iodide propyl iodide sodium chloride

Hydrolysis. The scission (or cleavage) of a molecule by reaction with water, with insertion of the elements of water into the final products, is called hydrolysis. An example is the acid-catalyzed hydrolysis of ethyl acetate:

$$CH_3COOC_2H_5 + H_2O \overset{H^+}{\rightleftharpoons} CH_3COOH + HOC_2H_5.$$

ethyl acetate water acetic acid ethyl alcohol

This reaction is typical of reversible reactions that do not go to completion. When one mole (the quantity with a weight in grams numerically equal to the molecular weight) of ethyl acetate and one mole of water react, only about one-third of the ethyl acetate is converted to acetic acid and ethyl alcohol. A greater degree of conversion may be produced by using an excess of water. However, a more practical procedure is to promote the reaction by a strong base, which reacts irreversibly with the acetic acid formed and, therefore, forces the reaction to completion.

Pyrolysis. Pyrolysis is the decomposition of a molecule by heat. The pyrolysis of carbon tetrachloride is a good example:

$$2CCl_4 \xrightarrow{800-900^\circ C} Cl_2C=CCl_2 + 2Cl_2\,.$$

carbon tetrachloride tetrachloroethylene chlorine

Condensation. The formation of a carbon–carbon bond between two molecules, with the elimination of a small molecule such as water or alcohol, generally is called a condensation reaction. An example follows:

$$CH_3COOC_2H_5 + CH_3COOC_2H_5$$

ethyl acetate ethyl acetate

$$\overset{^-OC_2H_5}{\rightleftharpoons} CH_3COCH_2COOC_2H_5 + C_2H_5OH.$$

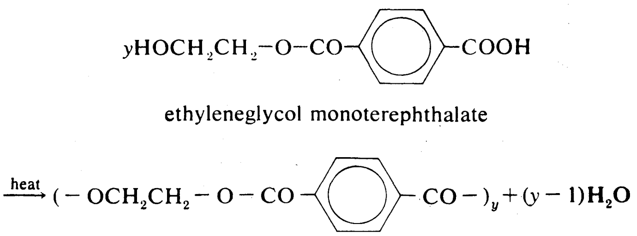
ethyl acetoacetate

Polymerization. Polymerization refers to the formation of compounds of high molecular weight—that is, polymers—from many smaller molecules (monomers) by addition or condensation reactions. Examples of both types are given below:

$$xCH_2=CH_2 \xrightarrow[\text{titanium tetrachloride, trialkyl-aluminum}]{TiCl_4,\ AlR_3} (-CH_2CH_2-)_x$$

ethylene polyethylene

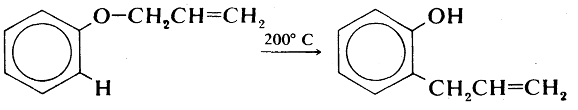
Condensation polymerization:



ethyleneglycol monoterephthalate

Dacron $y = 80$ to 130
molecular weight = 15,000 to 25,000

Molecular rearrangement. Molecular rearrangements are redistributions of atoms or groups within molecules. An example is the heat-induced rearrangement of allyl phenyl ether to 2-allylphenol, as shown:

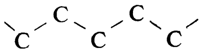


allyl phenyl ether 2-allylphenol

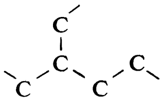
Numerous subclassifications of the above reactions, such as alkylation, acylation, esterification, ammonolysis, oxidation–reduction, nitration, chlorination, sulfonation, and disproportionation, also are used by chemists. See also the article CHEMICAL REACTIONS. (C.R.No./Ed.)

Hydrocarbons

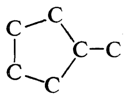
Chemical substances containing only the elements hydrogen (H) and carbon (C) generally are called hydrocarbons. In hydrocarbon molecules the carbon atoms make up the skeleton or framework that gives the molecules their general shape, and the hydrogen atoms are attached to the carbon atoms, completing the molecular structure. The carbon atoms may be arranged in chains that are consecutive or branched, in rings, or in complex three-dimensional frameworks.



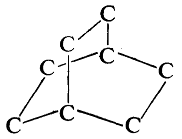
consecutive chain



branched chain



ring with one branch



a three-dimensional framework

The number of carbon atoms per molecule has no apparent upper limit: hydrocarbons with as many as 110 carbon atoms in a consecutive chain have been synthesized.

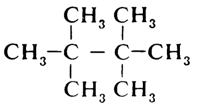
Occurrence. Many hydrocarbons occur in nature. As constituents of petroleum and natural gas, they are important energy sources. Because petroleum and natural gas are available in large supply, these hydrocarbons also serve as raw materials for the manufacture of plastics, synthetic rubber, solvents, explosives, lubricants, alcohols, synthetic fabrics, and many other useful products.

Certain hydrocarbons are present in trees and other plants. A hydrocarbon known as α-pinene (alpha-pinene), for example, with the empirical formula $C_{10}H_{16}$ (that is, with molecules each composed of ten carbon atoms and 16 hydrogen atoms), is the major component of turpentine, which is obtained from pine trees; and n-heptane, with the formula C_7H_{16}, can be obtained in pure form and in large

Hydrocarbons in foods, tobacco, and rubber

quantities from the Jeffrey pine. Hydrocarbons also are present in plant and insect waxes: beeswax contains 10–14 percent hydrocarbons (mainly $C_{31}H_{64}$).Hydrocarbons with fairly long carbon chains (up to C_{29}) have been isolated from cabbage and tobacco leaves, and from the cuticle (waxy film) of apples. Lycopene ($C_{40}H_{56}$) is the red pigment in ripe tomatoes and watermelon, and the carotenes (also with the formula $C_{40}H_{56}$ but with different arrangements of the atoms in the molecule) are important pigments in carrots and green leaves. Over 98 percent of crude plantation rubber is a hydrocarbon polymer, a chainlike molecule made of many small units joined together, with the formula $(C_5H_8)_n$, with n being a number averaging approximately 20,000.

Chemical and physical properties. Regardless of their diverse molecular structures, all hydrocarbons have a number of properties in common. They are virtually insoluble in water and, being less dense, float on its surface. Hydrocarbons are, however, usually soluble in one another as well as in certain organic solvents such as ether. Hydrocarbons may be gases, liquids, or solids at room temperature, their physical state depending mainly on their molecular weight. Those with four carbon atoms or fewer are gases. It is more difficult, however, to generalize about the demarcation between liquids and solids because crystallinity depends on molecular shape as well as on molecular weight. Some hydrocarbons, despite comparatively low molecular weights, are crystalline because their structures are highly symmetrical or compact. An example is hexamethylethane, the only hydrocarbon out of a total of 18, all of which have the formula C_8H_{18}, that is a solid at room temperature. It has the remarkably high melting point of 104° C (219° F), only 3° C (5.4° F) below its boiling point.



$$(CH_3)_3C-C(CH_3)_3$$

hexamethylethane

All hydrocarbons are combustible. If burned completely with sufficient oxygen, they produce carbon dioxide (CO_2) and water (H_2O). For example, the combustion of methane, the main component of natural gas, can be expressed:

$$\underset{\text{methane}}{CH_4} + \underset{\text{oxygen}}{2O_2} \longrightarrow \underset{\text{carbon dioxide}}{CO_2} + \underset{\text{water}}{2H_2O} + \text{heat}\,.$$

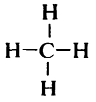
Such reactions are exothermic (that is, they give off heat), a fact that accounts for the widespread use of hydrocarbons as fuels. If insufficient oxygen is present, the hydrocarbon may be incompletely oxidized, producing carbon monoxide or carbon. Such incomplete combustion explains the presence of carbon monoxide in automobile exhaust, and carbon in the form of soot in chimneys.

$$\underset{\text{methane}}{CH_4} + \underset{\text{oxygen}}{\tfrac{3}{2}O_2} \longrightarrow \underset{\text{carbon monoxide}}{CO} + \underset{\text{water}}{2H_2O}$$

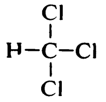
$$\underset{\text{methane}}{CH_4} + \underset{\text{oxygen}}{O_2} \longrightarrow \underset{\text{carbon}}{C} + \underset{\text{water}}{2H_2O}$$

Varieties of carbon–carbon bonds

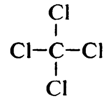
Classification. Hydrocarbons are in a sense the parent compounds from which can be derived all other organic compounds, as the large class of compounds containing carbon is called. In addition, the hydrocarbons illustrate several types of chemical bonding between carbon atoms, and their nomenclature forms the basis for the names of other organic compounds. More complicated molecules can be related structurally in a systematic way to hydrocarbons by the replacement of one or more of the hydrogens by other atoms or groups of atoms. Some examples are:



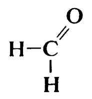
methane



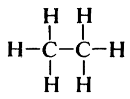
chloroform or trichloromethane



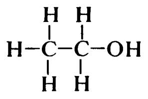
carbon tetrachloride or tetrachloromethane



formaldehyde or methanal

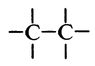


ethane

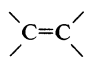


ethyl alcohol or ethanol

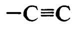
The structures of particular hydrocarbons and their chemistry depend in large measure upon the types of chemical bonds that the carbon atoms of the hydrocarbon molecules form with one another. Carbon atoms have a maximum valence, or chemical binding power, of four; this is a consequence of the fact that each carbon atom has four valence electrons, which it can share with other atoms including, of course, other carbon atoms. A single carbon atom, however, can share more than one of these electrons with another atom, again including carbon, thereby forming double or triple bonds, as shown in the following diagram. Thus, carbon atoms may be attached to four, three, or two other atoms.



single

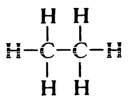


double

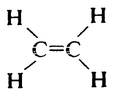


triple

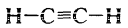
If each of the remaining bonds in the structures shown above is to a hydrogen atom, the structures of the following hydrocarbons are obtained:



ethane



ethylene



acetylene

In each case there are four bonds to each carbon atom and one bond to each hydrogen atom, hydrogen atoms always having a valence of one. The type of bonding present, whether single or multiple, affects the molecular structure and chemical behaviour of the hydrocarbon in characteristic ways, which will be discussed as each class of hydrocarbons is considered below.

In order to group together compounds with similar chemical behaviour, hydrocarbons are divided into several major classes according to their structure:

- Aliphatic hydrocarbons
  - Saturated (alkanes, cycloalkanes)
  - Unsaturated
    - Alkenes, dienes, polyenes; related cycloalkenes, etc.
    - Alkynes, cycloalkynes
- Aromatic hydrocarbons
  - Benzenoid (mono-, di-, and polycyclic)
  - Nonbenzenoid

The two major categories are aliphatic (Greek *aleiphar,* "fat") and aromatic (Greek *arōmatikos,* "aromatic"). The aliphatic hydrocarbons can be defined most simply as all those hydrocarbons that are not aromatic. The chemical sense of the word aromatic no longer has any relation to the term aroma from which it was originally derived. Rather, aromatic compounds are now loosely defined as those substances the molecules of which have closed rings of atoms joined by a type of bond regarded as a hybrid of the single bond and the double bond. (For further treatment of this special bond, see below *Aromatic hydrocarbons.*) Aliphatic substances include all hydrocarbons of which the molecules contain chains but no rings of carbon atoms, called acyclic, as well as those with carbon rings, called alicyclic, or carbocyclic, except for those classed as aromatic. Aliphatic hydrocarbons may be further cat-

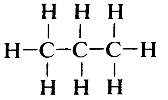
Saturated and unsaturated hydrocarbons

egorized according to the types of carbon–carbon bonds present. If all the bonds are single, the compound is said to be saturated. The systematic names for such classes of compounds are alkanes and cycloalkanes (or cyclanes). If a multiple bond connects any two carbon atoms, the hydrocarbon is referred to as unsaturated; the bonds may be double, as in alkenes or alkadienes, or triple, as in alkynes; some substances contain both types of multiple bonds in the same molecule. Aromatic compounds are classed as benzenoid if they contain six-membered aromatic rings, as in the parent compound benzene, or nonbenzenoid if they do not. They are further classified according to the number of rings they contain.

The nomenclature, structure, physical properties, synthesis, and chemical reactions of each of these classes of hydrocarbons are discussed below.

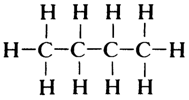
ALIPHATIC HYDROCARBONS

Alkanes. Alkanes have the general formula C_nH_{2n+2}, n being an integer and signifying the number of carbon atoms; each member of the class differs from the next higher and lower member, called homologues, by one carbon and two hydrogens. Formulas for the first two members of the series, methane, CH_4, and ethane, C_2H_6, have been given above. Only one structure is possible for each of these, and for the compound propane, C_3H_8, in which $n = 3$, but when $n = 4$, giving butane, C_4H_{10}, two structural formulas consistent with the valence rules can be written, depending on whether the carbon chain is consecutive or branched. In fact, substances corresponding to each butane structure can be prepared, as shown below. (The prefix "*n*-" refers to a normal, or consecutive, chain of atoms; "iso-" in this case implies a branched structure.)



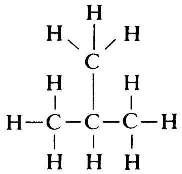
$H-\overset{H}{\underset{H}{C}}-\overset{H}{\underset{H}{C}}-\overset{H}{\underset{H}{C}}-H$

propane, C_3H_8



$H-\overset{H}{\underset{H}{C}}-\overset{H}{\underset{H}{C}}-\overset{H}{\underset{H}{C}}-\overset{H}{\underset{H}{C}}-H$

n-butane, C_4H_{10}



isobutane, C_4H_{10}

This phenomenon—called isomerism—is a general one, perhaps one of the most important in organic chemistry. Molecules are said to be isomers if they have the same molecular formula, as in the case of the butanes, C_4H_{10}, but different structural formulas (*i.e.*, different arrangements of the atoms in the molecule, as in *n*-butane and isobutane). Isomers usually have similar but not identical properties; both butanes, for example, are colourless gases, but *n*-butane boils at a slightly higher temperature (−0.5° C [32.9° F]) than does isobutane (−12° C [10.4° F]).

The number of possible isomers rises sharply with increasing *n*, as shown in Table 12. All alkanes up to and including the 75 isomers corresponding to formula $C_{10}H_{22}$ have been synthesized, and in each case the number of isomers found corresponded with that predicted. Of the groups containing more than 10 carbons, only a few isomers have been prepared, chiefly because there has been no practical incentive for chemists to synthesize them.

Table 12: Number of Possible Isomeric Alkanes

n	number of isomers	*n*	number of isomers
5	3	10	75
6	5	15	4,347
7	9	20	366,319
8	18	30	4,111,846,763
9	35	40	6.25×10^{13}

IUPAC nomenclature

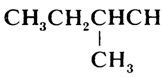
Nomenclature. Because of the existence of numerous isomers corresponding to each formula, the need for a systematic way of naming hydrocarbons quickly became apparent to chemists. A system originally devised by an international congress of chemists held in 1892 in Geneva, Switzerland, and later revised by the International Union of Pure and Applied Chemistry (IUPAC), is now used universally by organic chemists. In this system the saturated hydrocarbons are known as alkanes, the "*-ane*" ending being reserved for saturated compounds. After the first four members of the series, whose names have already been given, the normal (straight or consecutive chain) alkanes are named using Greek or Latin prefixes to indicate the number of carbon atoms in the chain:

C_5 *pent*ane	C_9 *non*ane	C_{13} *tridec*ane
C_6 *hex*ane	C_{10} *dec*ane	C_{14} *tetradec*ane
C_7 *hept*ane	C_{11} *undec*ane	C_{15} *pentadec*ane
C_8 *oct*ane	C_{12} *dodec*ane	C_{20} *eicos*ane

Branched-chain hydrocarbons are regarded as derivatives of normal hydrocarbons, the longest consecutive chain being considered the parent. The carbon atoms of this longest chain are then numbered from the end that gives the branches the lowest possible numbers. The names of the branches are attached as prefixes to the parent name, and their position is indicated by giving the number of the atom to which they are attached. If there are two branches on the same carbon, the number is repeated. The numbers precede the groups and are separated from them by hyphens. Finally, groups are named by changing the "-ane" ending to "-yl." For example:

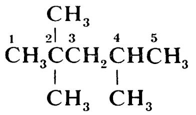
CH_4	CH_3-	CH_3CH_3	CH_3CH_2-
methane	methyl group	ethane	ethyl group

The system is illustrated by the following examples:



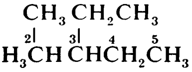
$CH_3CH_2\underset{CH_3}{CH}CH_3$

2-methylbutane



$\overset{1}{C}H_3\underset{CH_3}{\overset{CH_3}{\overset{2}{C}}}\overset{3}{C}H_2\underset{CH_3}{\overset{4}{C}}H\overset{5}{C}H_3$

2,2,4-trimethylpentane



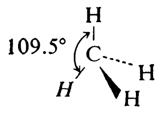
$\overset{1}{C}H_3\overset{CH_3}{\overset{2}{C}}H\,\overset{CH_2CH_3}{\overset{3}{C}}H\overset{4}{C}H_2\overset{5}{C}H_3$

3-ethyl-2-methylpentane
(alphabetic order for substituents)

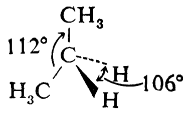
Special rules have been provided to handle unusually complex structures.

Tetrahedral structure of the carbon atom

Three-dimensional structures. Although flat, or two-dimensional, formulas are customarily used to represent the structures of organic compounds, the molecules themselves are in fact three-dimensional. The electron pairs making up chemical bonds usually adopt positions that keep them as far from one another as possible, since electrons carry like charges and repel one another. When the carbon atom has four groups attached to it (all with single bonds), the most stable arrangement is for the atom itself to be at the centre of a regular tetrahedron with the bonds extending toward its corners. Thus in methane, all H—C—H bond angles are equal to 109.5° (if the methane molecule were flat, the average angle between bonds would be only 90°.) The tetrahedral structure, with slight angle variations (106–112°) if the tetrahedron is not entirely symmetric, is common to all alkanes.



methane

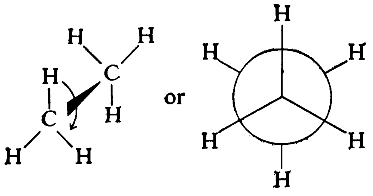


propane

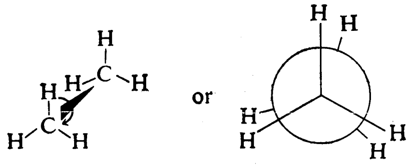
The carbon–carbon bond distance in alkanes is almost always 1.54 angstroms, and the carbon–hydrogen bond distance is 1.09 Å. The distance between hydrogens on adjacent carbons varies, however, because of rotation of the atoms about the carbon–carbon single bond. This phenomenon can be observed in the structure of ethane, in which the hydrogen–hydrogen distances are greatest in the so-called staggered conformation—that is, the shape the molecule assumes by rotation about the carbon–carbon bond; conversely, the hydrogen–hydrogen distances are smallest in another conformation, called the eclipsed conformation. Because of energy considerations, the staggered arrangement is more stable, and at room temperature most ethane molecules have this conformation.

Conformational analysis

Conformational analysis, the analysis of physical properties and chemical reactions as a function of molecular shape, was first developed during the 1950s by D.H.R. Barton of England and Odd Hassel of Norway, who shared a Nobel Prize in 1969 for their work.



ethane
(staggered conformation)

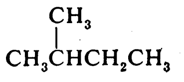


ethane
(eclipsed conformation)

Physical properties. The boiling points, melting points, and specific gravities of the normal alkanes increase with increasing molecular weight (Table 13). There is an enhanced attractive force between molecules as the number of atoms making them up increases, and this force affects melting and boiling points by making it more difficult to separate individual molecules (a prerequisite for either melting or boiling), and it increases specific gravities by pulling the molecules closer together. The viscosity, or "stickiness," also increases with increasing chain length, partly because of the same attractive force, and also because the greater lengths of the larger molecules increases the possibilities for entanglement and decreases the ability of the molecules to slip past one another. Branched-chain isomers almost always have lower boiling points than do the straight-chain isomers of the same compounds. This is clearly demonstrated by the three isomers of the hydrocarbon C_5H_{12} as shown below:

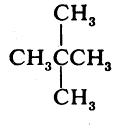
$CH_3CH_2CH_2CH_2CH_3$

pentane, 36.1° C



$$CH_3\overset{\displaystyle CH_3}{\overset{|}{C}}HCH_2CH_3$$

isopentane
or 2-methylbutane,
28° C



$$CH_3\underset{\displaystyle CH_3}{\underset{|}{\overset{\displaystyle CH_3}{\overset{|}{C}}}}CH_3$$

neopentane
or 2,2-dimethylpropane,
9.5° C

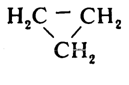
This effect stems from the fact that branching forces the molecules to remain farther apart and consequently decreases their mutual attraction.

Cycloalkanes. The general formula for cycloalkanes with one ring is C_nH_{2n}, where n is an integer equal to or greater than three. Thus, there are two fewer hydrogen atoms in the molecule of a cycloalkane than there are in the molecule of the corresponding alkane.

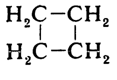
This condition exists because a valence bond from each of two carbon atoms is used in forming the ring. Rings are classified generally as small ($n = 3$ or 4), ordinary ($n = 5$ or 6), medium ($n = 7$ to 10), and large (n greater than 10). They are named with the same roots as alkanes, but with the prefix "cyclo-":

Table 13: Physical Properties of *n*-Alkanes

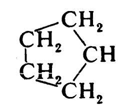
name	formula	boiling point (°C)	melting point (°C)	specific gravity
Methane	CH_4	−164	−182.5	.466
Ethane	C_2H_6	−88.6	−183.3	.572
Propane	C_3H_8	−42	−189.7	.501
Butane	C_4H_{10}	−0.5	−138.35	.601
Pentane	C_5H_{12}	+36.1	−129.7	.626
Hexane	C_6H_{14}	+68.9	−95.0	.660
Heptane	C_7H_{16}	+98.4	−90.6	.684
Octane	C_8H_{18}	+125.6	−56.8	.703
Nonane	C_9H_{20}	+150.8	−51.0	.718
Decane	$C_{10}H_{22}$	+174.1	−29.7	.730
Pentadecane	$C_{15}H_{32}$	+270	+10	.769
Octadecane	$C_{18}H_{38}$	+316.1	+28.2	.777
Eicosane	$C_{20}H_{42}$	+343	+36.8	.778
Triacontane	$C_{30}H_{62}$	+449.7	+65.8	.775
Tetracontane	$C_{40}H_{82}$	—	+81	—
Pentacontane	$C_{50}H_{102}$	—	+92	.794



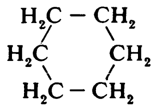
cyclopropane
(bp −33° C)



cyclobutane
(bp +12° C)



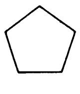
cyclopentane
(bp + 49.3° C)



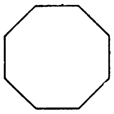
cyclohexane
(bp +80.7° C)

Frequently the formulas for cycloalkanes are represented by geometric figures that correspond to the rings. Each corner of the figure stands for a carbon atom with the appropriate number of hydrogen atoms to complete the valence of four. If substituents other than carbon are present they are shown explicitly, and the ring is numbered to assign them the lowest possible values. The "-ane" ending is retained when no multiple carbon–carbon bonds are present.

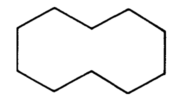
cyclopropane



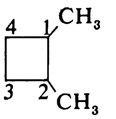
cyclopentane



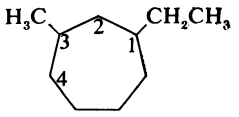
cyclooctane
(bp 147° C)



cyclodecane
(bp 201° C)

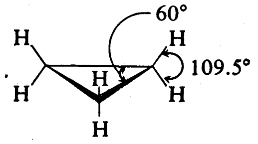


1,2-dimethylcyclobutane

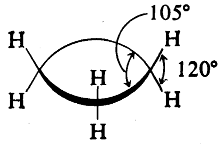


1-ethyl-3-methylcycloheptane

The three carbon atoms of cyclopropane necessarily define a plane, and the six hydrogen atoms are oriented above and below this plane. Since the carbons are at the corners of an equilateral triangle, the angles between the carbon–carbon bonds should be 60°. This small angle imposes considerable strain on the molecule, however, because the normal angle between carbon–carbon bonds (with the tetrahedral structure) is 109.5°. The molecule accommodates to this situation in two ways: (1) the angles between carbon–hydrogen bonds spread to 120° (larger than usual), allowing the angles between carbon–carbon bonds to contract, and (2) the electrons that constitute the carbon–carbon bonds do not lie along the direct lines



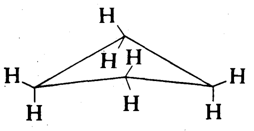
equilateral triangle
structure for cyclopropane



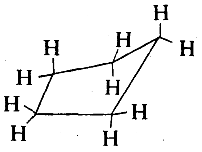
bent bond structure
for cyclopropane

between the centres of the carbon atoms, but rather they bulge outward in the plane of the ring, so that the actual bond angle becomes 105°. The chemical consequence of this "bending" of the bonds is that the electrons of the bond are more accessible than usual to attacking reagents. The cyclopropane ring, therefore, can be broken more easily than can larger rings.

Cyclobutane molecules also are strained. If the molecule were planar, each angle between carbon–carbon bonds would be 90°, and all hydrogens on adjacent carbons would be eclipsed. The ring puckers to decrease the interaction between hydrogens, even at the expense of some further decrease in the carbon–carbon bond angles. A puckered conformation is also adopted by cyclopentane, even though the internal angles of a regular pentagon are very close (108°) to the tetrahedral angle preferred by saturated carbon atoms. The puckered conformation in this case is assumed almost solely to avoid eclipsing of hydrogens.



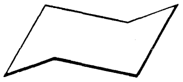
cyclobutane



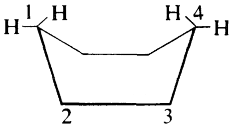
cyclopentane

Cyclohexane conformations

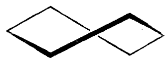
The conformations of cyclohexane, in particular, have been carefully studied because of the common occurrence of six-membered rings in many natural products, especially the steroids (*e.g.,* cholesterol). Three distinct conformations can be defined, the two rigid forms, chair and boat, and the flexible skew, or twist, form. In each, the angle between carbon–carbon bonds is 109.5° and strainless, and the forms are interconvertible by simple rotations about the C—C bonds. Since the angles are identical, the three conformations might be expected to be equally stable, but this is not the case. The chair conformation is much more stable, for in it all the hydrogen atoms on adjacent carbons are completely staggered, as in the staggered conformation of ethane. In the skew form several hydrogens are partially eclipsed, and in the boat form, which is the least stable, there are eclipsed hydrogens on C_2 and C_3, and again on C_5 and C_6, and in addition two of the hydrogens on C_1 and C_4 point in toward one another and are badly crowded. Because of these interactions, 999 of every 1,000 cyclohexane molecules will be in the chair form (the other being skew) at room temperature.



chair

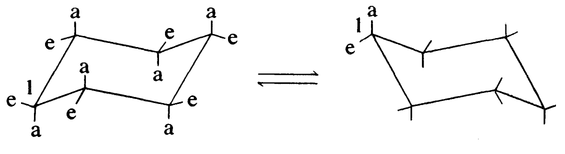


boat



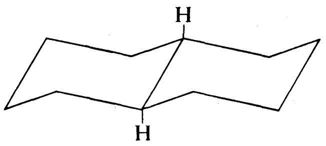
skew (twist)

There are two types of carbon–hydrogen bonds in the chair conformation of cyclohexane; these are called axial (*a*) or equatorial (*e*), depending on whether they are perpendicular or parallel to the mean molecular plane. The three axial hydrogens on each side of the molecule come close to one another in space. The two types (axial and equatorial) are interchangeable if the ring "flips" (this cannot be done directly; the skew forms are intermediates in the process). Any substituent other than hydrogen, the smallest atom, prefers an equatorial position since it avoids interaction with the axial hydrogens, and where "flipping" is possible this is the conformation observed. Conformational effects are important in explaining subtle differences in the chemical behaviour of cyclic molecules.

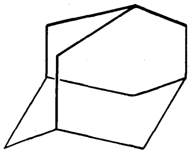


All larger rings also are puckered and free of angle strain, but those with seven to 12 carbons suffer unfavourable interactions between nonbonded hydrogens (hydrogens on different carbon atoms). The molecules adopt conformations that minimize these strains. Rings with more than 12 carbons are sufficiently flexible to adopt essentially strain-free conformations, analogous to those of the open-chain alkanes.

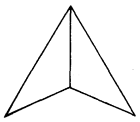
Many hydrocarbons with more than one ring are known. Among these are *trans*-decalin and adamantane, both having essentially strain-free molecules and both being substances found in petroleum. Three highly strained small-ring structures, bicyclobutane, spiropentane, and cubane, are examples of the unusual cyclic hydrocarbon molecules created by organic chemists in their laboratories.



trans-decalin, $C_{10}H_{18}$

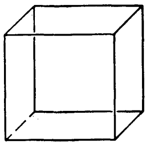


adamantane, $C_{10}H_{16}$



bicyclobutane, C_4H_6

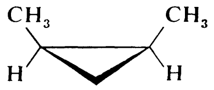
spiropentane, C_5H_8



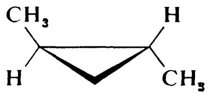
cubane, C_8H_8

Geometric isomerism

Certain cyclic structures show a type of isomerism in which two molecules differ only in the arrangement of substituent groups in space. For example, in 1,2-dimethylcyclopropane, the methyl groups may be on the same (*cis*) or on opposite (*trans*) sides of the ring plane. The resulting two substances are different, each having its own properties. *Cis–trans* isomers normally cannot be interconverted at room temperature, because to do so requires the breaking and remaking of chemical bonds. Alkenes, incidentally, also exhibit *cis–trans* isomerism (see below *Alkenes*). Biological reactions, controlled by enzymes, frequently are so sensitive to molecular structure that they can readily distinguish between a pair of geometric (*cis–trans*) isomers.



cis-1,2-dimethylcyclopropane
(bp 37°C)



trans-1,2-dimethylcyclopropane
(bp 29°C)

Sources and synthesis. The main commercial sources of alkanes are petroleum and natural gas. The latter contains 60–80 percent methane, 5–9 percent ethane, 3–18 percent propane, and 2–14 percent higher hydrocarbons. More than 150 pure hydrocarbons have been isolated from petroleum; about half are alkanes or cycloalkanes (the latter with C_5 or C_6 rings). Thermal and catalytic cracking (see below *Chemical reactions*) of petroleum and hydrogenation of coal are other alkane sources. Lower members of the series (through C_4) can be obtained in pure form by distillation. Annual production of each lower alkane is in the billion-pound range.

Individual higher alkanes and cycloalkanes usually are synthesized by reactions designed to give a unique product, since the number of possible isomers corresponding to each molecular formula is so large that separation of mixtures becomes impractical. One of the oldest but still useful laboratory methods involves electrolysis of organic acid salts (that is, passage of an electric current through a solution of the salt). Carbon dioxide is lost in the process, and the resulting organic radicals combine in pairs to form an alkane. For example,

$$2CH_3CH_2CH_2CO_2^-Na^+ - 2e$$

sodium butanoate

$$\xrightarrow[H_2O]{\text{electrolysis}} CH_3CH_2CH_2CH_2CH_2CH_3 + 2CO_2 + 2Na^+.$$

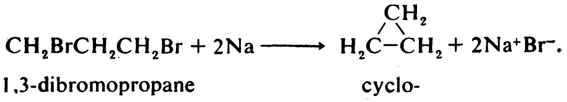
hexane

The reaction is quite general, the structure of the product depending upon the particular acid salt used. In a formally related reaction, the French chemist Charles-Adolphe Wurtz found in 1855 that organic halides couple when treated with sodium:

$$2CH_3CH_2CH_2Br + 2Na \longrightarrow CH_3CH_2CH_2CH_2CH_2CH_3 + 2Na^+Br^- .$$

1-bromopropane hexane

A modification of the Wurtz reaction has been used to prepare cyclopropane:



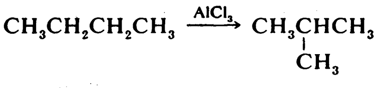
1,3-dibromopropane cyclopropane

In the current commercial process, zinc replaces sodium, and the cheaper 1-bromo-3-chloropropane is used. The product, cyclopropane, is a gas used in medicine as a general anesthetic.

Saturated hydrocarbons are also synthesized from corresponding unsaturated molecules, by the process of hydrogenation (see below *Alkenes; physical and chemical properties*).

Chemical reactions. Saturated hydrocarbons are unaffected at room temperature by most acids, alkalies, or oxidizing or reducing agents. This relative inertness gave rise to the name paraffins (Latin *parum,* "little"; *affinitas,* "affinity") sometimes applied to these substances, but it is now known that alkanes and cycloalkanes are more reactive than was once supposed. Methane, for example, will exchange its hydrogens for deuterium (a heavy isotope of hydrogen with chemical symbol, D) when placed in the extremely strong acid, deuterofluorosulfuric acid–antimony pentafluoride, FSO_3 D–SbF_5, at 80° C (176° F). Also, certain alkanes can be isomerized to highly branched structures. This isomerization reaction, carried out with the aid of an aluminum chloride catalyst, is used commercially to convert *n*-butane to isobutane, which is an important raw material for high-octane fuels.

Isomerization reactions



$$CH_3CH_2CH_2CH_3 \xrightarrow{AlCl_3} CH_3\underset{\substack{|\\ CH_3}}{C}HCH_3$$

n-butane isobutane

The most common reactions of saturated hydrocarbons involve replacement of one or more of the hydrogens by other atoms or groups of atoms. The reaction was discovered in a curious way. At a ball in France in 1830 guests were driven from the ballroom by choking fumes given off by the burning candles. It was found that the candles had been bleached by a process using chlorine; some hydrogen atoms in the hydrocarbon molecules of the wax were replaced by chlorine atoms and, on heating gave off hydrogen chloride fumes. The reaction was investigated systematically and found to be useful synthetically. When an alkane, or cycloalkane, is treated with chlorine or bromine at elevated temperature or in the presence of ultraviolet light, substitution occurs in the following manner:

$$RH + Cl_2 \xrightarrow[\text{light}]{\text{heat or}} RCl + HCl.$$

(R = an organic radical or group of atoms.) More than one hydrogen on a given carbon can be replaced, as in the sequence:

$$CH_4 \xrightarrow[-HCl]{Cl_2} CH_3Cl \xrightarrow[-HCl]{Cl_2} CH_2Cl_2 \xrightarrow[-HCl]{Cl_2} CHCl_3 \xrightarrow[-HCl]{Cl_2} CCl_4 .$$

methane methyl chloride dichloromethane chloroform carbon tetrachloride

The reaction conditions can be modified to give predominant mono- or poly- substitution, as desired.

Cracking

At high temperatures (500–700° C [932–1,292° F]) higher alkanes undergo rupture, or cracking, to give a mixture of smaller organic molecules and hydrogen gas. It is clear that if a given alkane, such as $C_{20}H_{42}$, is cracked to furnish another acyclic alkane, perhaps $C_{10}H_{22}$, along with a molecule of hydrogen (H_2), some of the remaining fragments must be unsaturated since, in the example chosen, only 18 hydrogen atoms are available to be distributed among the remaining ten carbon atoms. Cracking is therefore an important commercial source of unsaturated hydrocarbons.

Finally, of course, all alkanes are fuels, which can be burned to carbon dioxide and water. Partial oxidation occurs spontaneously at a considerably lower temperature than that required for combustion. This process, called autoxidation, causes the formation of organic acids and other corrosive materials during the use of hydrocarbon lubricating oils. Various antioxidants and inhibitors are added to automotive oils to slow down this undesirable reaction.

Alkenes. Alkenes, also called olefins, have two fewer hydrogens per molecule than do alkanes. They are unsaturated, with the general formula C_nH_{2n}, and their most characteristic structural feature is a carbon–carbon double bond. The first two members of the series are ethylene, C_2H_4, and

$CH_2{=}CH_2$	$CH_3CH{=}CH_2$
ethylene (ethene) (bp −102.4° C)	propylene (propene) (bp −47.7° C)

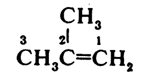
propylene, C_3H_6. The number of possible isomers increases more rapidly than with alkanes; whereas there are only two butanes, there are three isomeric butenes (see below *Geometry of the double bond,* however, for other, geometric isomers). The double bond may occupy any of several positions along the chain. If the bond is at the end of the chain, it is called terminal; otherwise, internal.

$$\overset{1}{C}H_2{=}\overset{2}{C}H\overset{3}{C}H_2\overset{4}{C}H_3$$

1-butene (bp −6.5° C)

$$\overset{1}{C}H_3\overset{2}{C}H{=}\overset{3}{C}H\overset{4}{C}H_3$$

2-butene



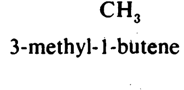
$$\overset{3}{C}H_3\overset{\substack{CH_3\\ 2|}}{C}{=}\overset{1}{C}H_2$$

isobutylene (2-methylpropene) (bp −6.6° C)

Nomenclature. The general name is alkene, the "-ene" ending standing for a carbon–carbon double bond. (Older names include an "-yl-" after the root, as in ethylene, but the IUPAC names omit this syllable, as in ethene.) The longest chain incorporating the double bond determines the root of the name, which is the same as that of the alkane with the same number of carbon atoms. The position of the double bond is designated by the lower number of the two carbons involved in the bond, and the chain is numbered so as to give the double bond the lowest possible number (as in the formulas of 1- and 2-butene, above). The number that refers to the position of the double bond is sometimes omitted if this causes no ambiguity, as is often the case with cycloalkenes. Substituents are named in the same way as with alkanes, the numbering being determined, however, by the position of the double bond. These rules are illustrated by the examples at the end of this section.

$$\overset{1}{C}H_3\overset{2}{C}H{=}\overset{3}{C}H\overset{4}{C}H_2\overset{5}{C}H_2\overset{6}{C}H_3$$

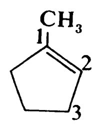
2-hexene



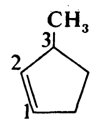
$$\overset{1}{C}H_2{=}\overset{2}{C}H\underset{\substack{|\\ CH_3}}{\overset{3}{C}}H\overset{4}{C}H_3$$

3-methyl-1-butene

cyclopropene



1-methyl-cyclopentene

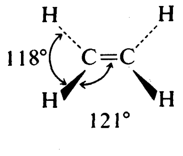


3-methyl-cyclopentene

Vinyl and allyl groups

Two unsaturated groups have useful common names, the vinyl and allyl groups. The vinyl group is $CH_2{=}CH-$, forming, *e.g.*, $CH_2{=}CHCl$, vinyl chloride, or chloroethene. The allyl group is $CH_2{=}CH-CH_2-$, forming $CH_2{=}CHCH_2Br$, allyl bromide, or 3-bromopropene.

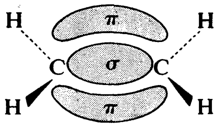
Geometry of the double bond. Each of the carbon atoms involved in a double bond is attached to only two other atoms or groups, because the total number of bonds to carbon can be no greater than its normal valence of four. In principle, the best geometric arrangement, which minimizes the repulsion between the electrons in the bonds, has all four atoms (carbon and the three atoms attached to it) in a single plane, with three equal bond angles of 120°. In practice the planar geometry is observed but the angles vary slightly from 120°, as shown for ethylene.



The average carbon–carbon double-bond distance is 1.34 Å, appreciably shorter (by 0.2 Å) than that of the carbon–carbon single bond. This decrease results from the fact that two electron pairs, rather than only one, bind the two carbon nuclei to one another.

Sigma and pi bonds

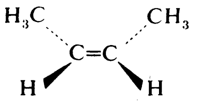
The position of the two electron pairs is of some significance in explaining the reactions of alkenes. One pair, as in a single bond, occupies the space directly between the two carbon nuclei, in an orbital (region of space within which the electrons are localized) that is symmetric about an axis drawn between the two nuclei. Bonds with this symmetry are referred to as σ (sigma) bonds. The second electron pair, in order that it not occupy the same space as the first, cannot lie directly between the nuclei that it binds together. It lies instead in an orbital that is both immediately above and immediately below the plane formed by the two carbons and the atoms attached to them; this is called a π (pi) bond. The π bond is somewhat weaker and more easily broken than the σ bond. The π electrons, not being buried between two atomic nuclei, are quite accessible to chemical reagents. For this reason, alkenes and cycloalkenes are much more reactive than alkanes, and most of the reactions involve breaking the π bond or adding to it.



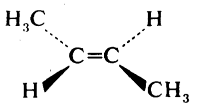
Cis and trans isomers

The presence of two bonds between adjacent carbon atoms imposes some geometric restrictions on the molecule. Sigma bonds, being centrosymmetric about the C-C axis, permit relatively free rotation of one carbon with respect to another, as in ethane. But the π bond has no such symmetry; to rotate one carbon of a double bond 180° with respect to the other carbon would require that the π bond be broken and eventually reformed. This process usually requires considerably more energy than is available at room temperature. Consequently, rotation about double bonds is ordinarily not possible except at high temperatures, or in the presence of other energy sources, or during chemical reactions in which the π bond is temporarily broken anyway. Restricted rotation around double bonds leads to geometric, or *cis–trans,* isomers, like those found in cycloalkanes (see above *Cycloalkanes*), when the two groups attached to each carbon of the double bond are different. Thus there are two 2-butenes, *cis* and *trans.* In all small and ordinary sized ring compounds with a double bond (cycloalkenes), the geometry is necessarily *cis;* the smallest stable ring structure with a *trans* double bond has eight carbon atoms.

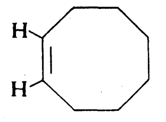
Geometric isomers may differ rather drastically in molecular shape. The carbon chain in *cis*-2-butene, for example, is approximately U-shaped, whereas that of the *trans* isomer is Z-shaped. In biological systems such drastic structural changes can be very significant, because they affect the way in which an unsaturated molecule interacts with enzymes. The chemistry of vision constitutes a particularly important example. A visual pigment called retinene contains a number of double bonds, and the conversion of one of these from the *cis* to the *trans* form occurs when the pigment is exposed to light. This isomerization alters the degree of attachment of retinene to a protein molecule and brings about other changes in the chemicals of the retina.



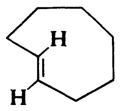
cis-2-butene
(bp +3.7° C)



trans-2-butene
(bp +0.88° C)



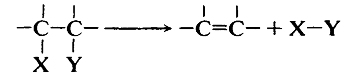
cis-cyclooctene



trans-cyclooctene

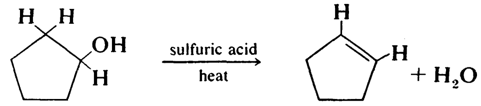
Sources and synthesis. The lower alkenes (through C_4) are produced commercially by cracking natural gas or petroleum or mixtures of hydrocarbons derived from them. The largest percentage of the ethylene produced each year in the United States is used as a raw material for making the plastic polyethylene. Another sizable percentage goes to make ethylene oxide, which in turn is used to manufacture glycol antifreeze and other products. Substantial amounts also are converted into ethyl alcohol. Propylene and the butenes, also manufactured on a large scale, are sources of other chemicals that are used as solvents or as starting materials for the manufacture of plastics, detergents, synthetic rubber, food preservatives, and numerous other products.

Individual higher alkenes and cycloalkenes may be synthesized by reactions in which a double bond is introduced into a saturated precursor. The principle behind most of these methods is expressed by the equation:

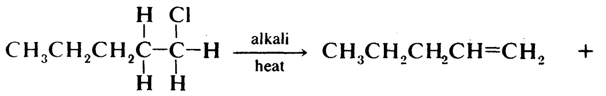


$$-\underset{X}{\overset{|}{\underset{|}{C}}}-\underset{Y}{\overset{|}{\underset{|}{C}}}- \longrightarrow -\overset{|}{C}{=}\overset{|}{C}- + X{-}Y$$

in which X and Y are two atoms or groups attached to adjacent carbon atoms and capable of forming some small molecule X—Y. The process is called an elimination reaction (X and Y being eliminated from the starting compound). Examples include the dehydration of alcohols and the removal of a halogen acid from an alkyl halide.



cyclopentanol, an alcohol — cyclopentene



$$CH_3CH_2CH_2\underset{H}{\overset{H}{C}}-\underset{H}{\overset{Cl}{C}}-H \xrightarrow[\text{heat}]{\text{alkali}} CH_3CH_2CH_2CH{=}CH_2 \quad + \quad HCl$$

an alkyl halide — an alkene — hydrochloric acid

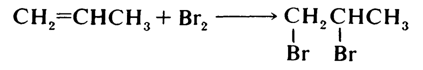
These usually are laboratory rather than commercial methods. Alkenes also can be prepared by partial hydrogenation of alkynes (see below *Alkynes; reactions*).

Physical and chemical properties. The physical properties of the alkenes are generally similar to those of alkanes or cycloalkanes with equal numbers of carbon atoms.

Addition reactions

In contrast with the alkanes, however, which react predominantly by substitution, alkenes react mainly by addition; various reagents add to the carbon–carbon double bond by reactions that are the reverse of elimination. These reactions are frequently very rapid, even at room temperature.

Halogens, such as bromine, add rapidly to most alkenes. The product is saturated since the carbon–carbon double bond in the reactant is replaced entirely by single bonds in the product. The π electrons originally present are used to form single (σ) bonds to the bromine atoms. One bromine adds to each carbon of the double bond (note that the only product from propene is 1,2-dibromopropane; no 1,1- or 1,3- or 2,2-isomer is formed).

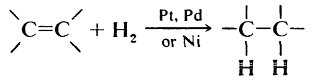


$$CH_2{=}CHCH_3 + Br_2 \longrightarrow \underset{Br}{C}H_2\underset{Br}{C}HCH_3$$

propene 1,2-dibromopropane

The reaction with bromine is frequently used as a test for unsaturation in a molecule. Alkenes are colourless, bromine is a reddish-brown liquid, and the product, an organic dibromide, is colourless. When a reddish-brown solution of bromine in some inert solvent (*e.g.,* carbon tetrachloride) is added to an alkene, the bromine reacts instantly to form the colourless product. The bromine solution is "decolourized." If the substance being tested has no carbon–carbon double bonds, the bromine does not react, or reacts only slowly by substitution, and the colour of the bromine solution remains essentially unchanged. The reaction can be modified and used to determine the percentage of alkenes present in an unknown mixture; bromine is added until it is no longer decolourized, and the amount required for a given sample weight is measured.

Alkenes can be converted to alkanes by hydrogen, usually under pressure and in the presence of a finely divided metal catalyst, such as platinum, palladium, or nickel. The generality of this reaction, called catalytic hydrogenation, was first appreciated by the French chemist Paul Sabatier (Nobel Prize, 1912). Hydrogenation is used commercially not only to make alkanes or cycloalkanes, but in general to produce more saturated molecules from less saturated ones (for example, oleomargarine from soybean oil).

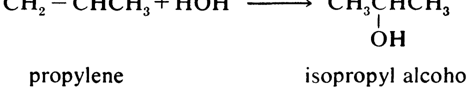


Alkenes are converted to alcohols by hydration; the reaction requires an acid catalyst. Most of the ethanol produced for industrial use is obtained from ethylene.

$$CH_2{=}CH_2 + HOH \xrightarrow[\text{acid}]{\text{sulfuric}} CH_3CH_2OH$$

ethyl alcohol
(ethanol)

The following similar reaction gives isopropyl alcohol from propylene.



$$CH_2 = CHCH_3 + HOH \xrightarrow{\text{acid}} CH_3\underset{OH}{C}HCH_3$$

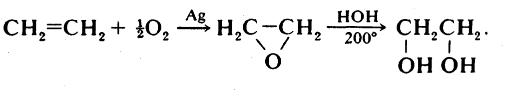
propylene isopropyl alcohol

Alcohols also add to alkenes in the presence of acid. Ether (more correctly, diethyl ether), the widely used solvent and anesthetic, is made from ethylene in this way.

$$CH_2{=}CH_2 + HOCH_2CH_3 \xrightarrow{\text{acid}} CH_3CH_2OCH_2CH_3$$

ethylene ethyl alcohol diethyl ether

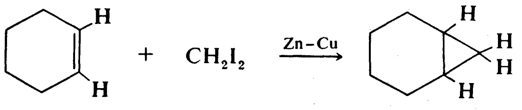
Ethylene oxide is produced by passing a mixture of ethylene and air (or oxygen) over a heated silver catalyst:



ethylene oxide (oxirane) ethylene glycol

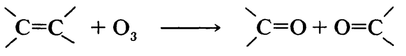
Most of the ethylene oxide produced is converted (with water) to ethylene glycol, the main component of permanent automotive antifreeze.

Alkenes may be converted to cyclopropanes by reaction with dihalomethanes (CH_2X_2) and a metal; in effect, a $-CH_2-$ group is added to the double bond, as for example:



diiodomethane norcarane

Alkenes and cycloalkenes react rapidly with ozone. Although ozone first adds to the double bond, the net reaction is cleavage of the molecule, a process called ozonolysis; two fragments with C=O (carbonyl) groups are produced.



ozone

When the position of the carbon–carbon double bond in the molecule is unknown, the ozonolysis reaction can be used to locate it, if the carbonyl products can be identified. For example, on ozonolysis 2-butene gives a single product, acetaldehyde, $CH_3CH{=}O$, whereas 1-butene gives two products, formaldehyde, $CH_2{=}O$, and propionaldehyde, $O{=}CHCH_2CH_3$.

Polymerization

One alkene molecule can add to the double bond of another; each is said to be a monomer, and the product is a dimer. If the process is repeated, trimers, and eventually polymers, substances composed of a great many monomer units, are obtained. Polymerization is one of the most important reactions used by the chemical industry.

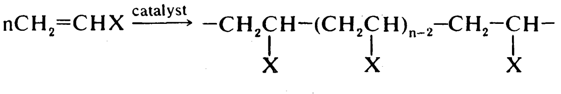
Ethylene can be converted to polyethylene by several processes:

$$nCH_2{=}CH_2 \xrightarrow{\text{catalyst}} -CH_2CH_2{-}(CH_2CH_2)_{n-2}{-}CH_2CH_2-$$

polyethylene

the product is essentially an alkane and therefore is chemically inert. The ends of the chain may have catalyst molecules attached, or the chain may terminate by loss of one hydrogen, leaving a double bond at the end. Depending on the catalyst used, the chain may be almost entirely consecutive as shown, or it may have occasional branches of smaller chains. The more nearly linear the chain, the greater the density of the polymer. Because of its high molecular weight (n often exceeds 1,000), polyethylene is insoluble in most solvents. But it is thermoplastic (softens and flows on heating), and it can be extruded into sheets or films and molded into various shapes that are retained on cooling.

Substituted ethylenes also form polymers according to the reaction

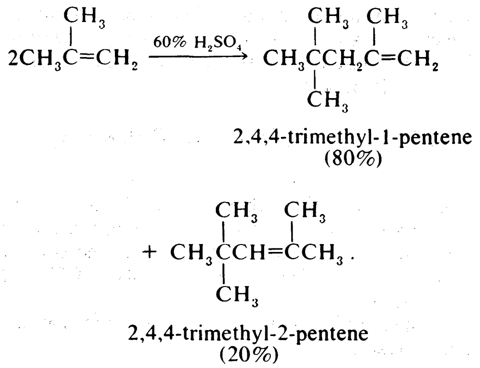


$$nCH_2{=}CHX \xrightarrow{\text{catalyst}} -CH_2\underset{X}{C}H{-}(CH_2\underset{X}{C}H)_{n-2}{-}CH_2{-}\underset{X}{C}H-$$

a vinyl compound a vinyl polymer

If X = Cl the product is polyvinyl chloride, over a billion pounds of which are used annually to make floor tiles, shoe soles, raincoats, phonograph records, and textile fibres. If X = C_6H_5 (a phenyl group, derived from benzene: see below *Aromatic hydrocarbons*), the product is polystyrene, often foamed and used as a lightweight structural and insulating material. If X = CH_3, the product is polypropylene, used to make films, molded articles, and fibres.

Some alkenes form useful dimers or other lower polymers. For example, if isobutylene is treated with aqueous sulfuric acid, the products are mainly the two dimers shown in the equation



$$2CH_3\overset{CH_3}{\overset{|}{C}}{=}CH_2 \xrightarrow{60\%\ H_2SO_4} CH_3\underset{CH_3}{\overset{CH_3}{C}}CH_2\overset{CH_3}{C}{=}CH_2$$

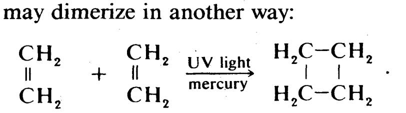
2,4,4-trimethyl-1-pentene
(80%)

$$+\ CH_3\underset{CH_3}{\overset{CH_3}{C}}CH{=}\overset{CH_3}{C}CH_3\,.$$

2,4,4-trimethyl-2-pentene
(20%)

Although isobutylene is a gas, with too low a boiling point for effective use as gasoline, the dimers fall nicely within the gasoline boiling range, and their highly branched structures are particularly desirable in preventing engine knock. Hydrogenation of the dimer mixture produces 2,2,-4-trimethylpentane, the hydrocarbon used as a standard for 100 octane number. Both reactions, dimerization and hydrogenation, are carried out in most petroleum refineries to produce high-octane gasolines.

In the presence of ultraviolet light and a sensitizer, alkenes may dimerize in another way:



$$\begin{matrix}CH_2\\ \|\\ CH_2\end{matrix} + \begin{matrix}CH_2\\ \|\\ CH_2\end{matrix} \xrightarrow[\text{mercury}]{\text{UV light}} \begin{matrix}H_2C-CH_2\\ |\qquad |\\ H_2C-CH_2\end{matrix}\,.$$

The reaction, called a cycloaddition, is useful for laboratory synthesis of cyclobutanes.

Alkynes. A carbon–carbon triple bond is the characteristic structural feature of alkynes, or acetylenes. The simplest and commercially most important of the series, acetylene itself, was first prepared as long ago as 1836, though its name was coined some 25 years later. The general formula for the class is C_nH_{2n-2}.

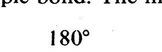
Nomenclature and structure. Alkynes are named in the same way as alkenes, except that the ending that designates the triple bond is "-yne." Acetylene is almost always called by that name, though its IUPAC name is ethyne:

$$HC{\equiv}CH \qquad CH_3C{\equiv}CH \qquad \overset{1}{C}H_3\overset{2}{C}{\equiv}\overset{3}{C}\overset{4}{C}H_3\,.$$

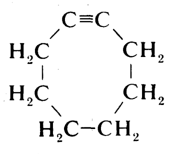
ethyne (acetylene)	propyne (methylacetylene)	2-butyne (dimethylacetylene)

Other alkynes usually are named systematically, but they also may be named as derivatives of acetylene (examples of these names are given above in parentheses).

Only one other atom or group can be attached to a carbon atom of a triple bond. The molecule is linear:



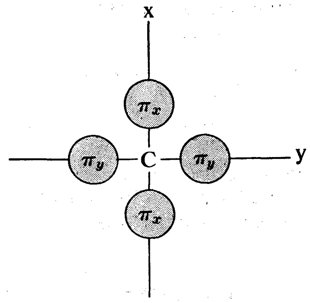
Because of this linearity, small rings incorporating a triple bond are not stable. Cyclooctyne is the smallest cyclic acetylene of which synthesis has been reported, though evidence has been obtained for smaller cycloalkynes as reaction intermediates (which are not actually isolated).



cyclooctyne

The C≡C distance is 1.20 Å, appreciably shorter than both the C═C and C—C distances (see above *Geometry of the double bond*), because three electron pairs bind the two carbon atoms in the first case. The triple bond consists of one sigma (σ) bond along the C—C internuclear axis, and two pi (π) bonds lying in mutually perpendicular planes. A view down the carbon–carbon axis (the z axis) shows only the π bonds, the σ bond being symmetric along the z axis. The two groups attached to the triple bond also lie along the z axis, so that no geometric (*cis–trans*) isomerism is possible.

The carbon–carbon triple bond



Physical properties. In general, alkynes boil 10°–20° C (18°–36° F) higher than do the corresponding alkenes and alkanes. They are slightly more dense than alkenes, and more soluble in water. Owing to its symmetric structure, acetylene has an unusually high melting point, within a few degrees of its boiling point (−83.4° C [−118.1° F]). When liquefied, acetylene is sensitive to shock and may explode, forming its elements: $C_2H_2 \rightarrow 2C + 2H_2$ + heat. For this reason the gas cannot be liquefied safely under pressure for shipping. Acetylene gas, however, is remarkably soluble in acetone, and at a pressure of 12 atmospheres (180 pounds per square inch) one volume of acetone will dissolve 300 volumes of acetylene. Since this solution is stable, acetylene usually is transported under pressure in tanks filled with porous material, such as asbestos, that is saturated with acetone.

Other alkynes, especially those with an internal triple bond, are less sensitive than is acetylene itself. Internal alkynes generally boil higher than the isomeric terminal alkynes, as shown by the last two pairs of entries in Table 14.

Commercial production of acetylene

Sources and synthesis. Acetylene is the only alkyne produced commercially in large amounts. It is made by the reaction of water with calcium carbide, which, in turn, comes from coke and lime, as represented in the following reactions:

Table 14: Boiling Points of Alkynes

name	carbon skeleton	boiling point (°C)
Acetylene	C≡C	−84.0
Propyne	C≡C—C	−23.2
1-Butyne	C≡C—C—C	+8.1
2-Butyne	C—C≡C—C	+27.0
1-Pentyne	C≡C—C—C—C	+40.2
2-Pentyne	C—C≡C—C—C	+56.1

$$3C + CaO \xrightarrow{2{,}500^\circ} CaC_2 + CO$$

coke	lime	calcium carbide	carbon monoxide

$$CaC_2 + 2H_2O \xrightarrow[\text{temperature}]{\text{room}} HC{\equiv}CH + Ca(OH)_2\,.$$

acetylene	calcium hydroxide

Acetylene also is a major product from high-temperature cracking of methane:

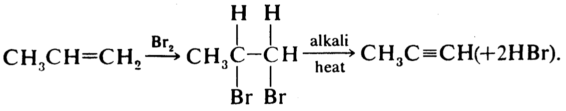
$$2CH_4 \xrightarrow{1{,}600^\circ} C_2H_2 + 3H_2;$$

a considerable fraction of commercial acetylene comes from this source.

Before the era of electricity, acetylene was used as a component of illuminating gas, for it burns with a colourless flame. Now about half the acetylene produced is consumed in oxyacetylene torches for welding and cutting metals; the oxyacetylene flame is intensely hot (about

2,800° C [5,000° F]). The remaining acetylene produced is used as a raw material for other organic chemicals (see below *Reactions*).

Other alkynes can be made from acetylene (see below *Reactions*) or by extensions of the methods used for introducing double bonds, as, for example:



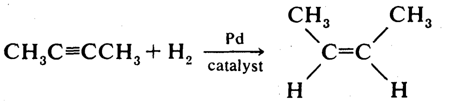
$$CH_3CH{=}CH_2 \xrightarrow{Br_2} CH_3\underset{Br}{\overset{H}{C}}{-}\underset{Br}{\overset{H}{C}}H \xrightarrow[\text{heat}]{\text{alkali}} CH_3C{\equiv}CH(+2HBr).$$

Reactions. Many of the same reagents that add to alkenes also add to alkynes, but because of the triple bond the latter are capable of adding two molecules of reagent rather than one:

$$HC\equiv CH \xrightarrow{Br_2} CHBr{=}CHBr \xrightarrow{Br_2} CHBr_2CHBr_2.$$

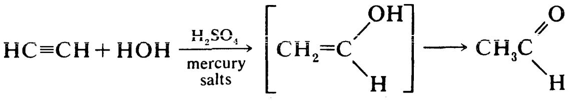
acetylene 1,2-dibromoethene 1,1,2,2-tetrabromoethane

When hydrogen and a special palladium catalyst are used, hydrogenation can be stopped readily at the alkene stage. Since both of the hydrogen atoms are delivered to the same side of the triple bond from the catalyst surface, this reaction provides a good method for obtaining *cis*-alkenes:



cis-2-butene

Perhaps the most important commercial reaction of acetylene is its hydration. The initial product (vinyl alcohol) is unstable and rearranges to the carbonyl compound acetaldehyde, an important precursor of many other industrial organic chemicals.



vinyl alcohol acetaldehyde

Hydrogen cyanide also adds to acetylene. The product, acrylonitrile, is a valuable monomer used to make synthetic fibres (Acrilan, Orlon, Dynel).

$$HC{\equiv}CH + HC{\equiv}N \xrightarrow[\text{heat}]{\text{catalyst}} CH_2{=}CHC{\equiv}N$$

acrylonitrile

Terminal alkynes are very weak acids, and the hydrogen attached to the triple bond may be replaced by metals, producing salts called acetylides, although the calcium salt is almost always called the carbide. This process is accomplished most commonly by reaction in liquid ammonia with sodium amide, a very strong base:

$$RC{\equiv}CH + Na^+NH_2^- \xrightarrow[\text{ammonia}]{\text{liquid}} RC{\equiv}C^-Na^+ + NH_3.$$

a sodium acetylide

Acetylides are reactive, and can be used to introduce triple bonds into other molecules. Sodium acetylide, for example, reacts with ethyl bromide, to produce 1-butyne, in a reaction somewhat reminiscent of the Wurtz reaction (see above *Alkanes; Sources and synthesis*):

$$HC\equiv C^-Na^+ + BrCH_3 \longrightarrow HC\equiv CCH_3 + Na^+Br^-$$

propyne

In principle any higher alkyne can be made from acetylene itself by judicious use of this reaction.

Silver, copper, and other heavy metal acetylides, when dry, explode under the influence of heat or shock, and hence they are used in detonators. In suspension, however, copper acetylides are useful in synthesis as, for example, in oxidative coupling of two alkynes to produce a product having two triple bonds:

$$2RC{\equiv}CH + \tfrac{1}{2}O_2 \xrightarrow[\text{salt}]{\text{cuprous}} RC{\equiv}C{-}C{\equiv}CR + H_2O.$$

Some hydrocarbons in which triple bonds are separated by single bonds have been found to be present in certain fungi and higher plants; an example is the substance $CH_3{-}C{\equiv}C{-}C{\equiv}C{-}C{\equiv}C{-}C{\equiv}C{-}C{\equiv}C{-}CH{=}CH_2$, present in Asteraceae.

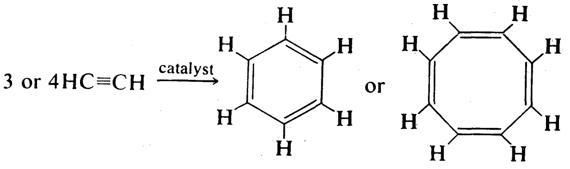
Acetylenes also undergo another type of dimerization, or self-addition, reaction:

Dimerization of acetylenes

$$HC{\equiv}CH + HC{\equiv}CH \xrightarrow[\substack{\text{ammonium chloride}\\ \text{hydrochloric acid}}]{\text{cuprous chloride}} CH_2{=}CHC{\equiv}CH.$$

vinylacetylene

Vinylacetylene is employed in the synthesis of neoprene, a synthetic rubber. With still other catalysts acetylenes undergo cyclic trimerization or tetramerization:



benzene cyclooctatetraene

In a strong base they may be isomerized to other alkynes or to compounds with two double bonds on the same carbon atom, allenes:

$$CH_3CH_2C{\equiv}CH \underset{\text{base}}{\overset{\text{strong}}{\rightleftharpoons}} CH_3CH{=}C{=}CH_2 \underset{\text{base}}{\overset{\text{strong}}{\rightleftharpoons}} CH_3C{\equiv}CCH_3.$$

1-butyne 1,2-butadiene (an allene) 2-butyne

From the variety of reactions they undergo, it is understandable that alkynes are particularly useful synthetic intermediates in organic chemistry.

Dienes and polyenes. Compounds with more than one carbon–carbon double bond occur widely in nature and are commercially important. The chemical properties of dienes, compounds with two C=C bonds, depend on the relative positions of the double bonds. The most important class, having alternate double and single bonds, are called conjugated dienes. Other classes have isolated double bonds (separated by two or more single bonds) or cumulative double bonds as in allenes.

Conjugated dienes

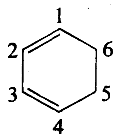
$C{=}C{-}C{=}C$	$C{=}C{-}C{-}C{=}C$	$C{=}C{=}C$
conjugated	isolated	cumulative

Nomenclature. The system for naming polyenes is the same as for alkenes, except that the positions of all double bonds must be designated by appropriate numbers. The following examples are illustrative:

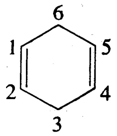
$CH_2{=}C{=}CH_2$ propadiene (allene)

$\overset{1}{C}H_2{=}\overset{2}{C}H\overset{3}{C}H{=}\overset{4}{C}H_2$ 1,3-butadiene

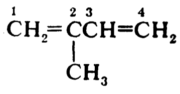
$\overset{1}{C}H_2{=}\overset{2}{C}{=}\overset{3}{C}H\overset{4}{C}H_3$ 1,2-butadiene



1,3-cyclohexadiene



1,4-cyclohexadiene



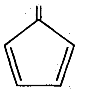
$$\overset{1}{C}H_2{=}\underset{\substack{|\\CH_3}}{\overset{2}{C}}\overset{3}{C}H{=}\overset{4}{C}H_2$$

2-methyl-1,3-butadiene (isoprene)

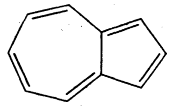
$\overset{1}{C}H_2{=}\overset{2}{C}H\overset{3}{C}H_2\overset{4}{C}H_2\overset{5}{C}H{=}\overset{6}{C}H_2$ 1,5-hexadiene

$\overset{1}{C}H_2{=}\overset{2}{C}H\overset{3}{C}H{=}\overset{4}{C}H\overset{5}{C}H{=}\overset{6}{C}H_2$ 1,3,5-hexatriene

Physical properties. The physical properties of polyenes, such as boiling point and density, are similar to those of alkanes or alkenes with an equal number of carbon atoms. The most exceptional characteristic of polyenes is the intense colour often associated with those in which the double bonds are conjugated. Some examples (with only the skeletons shown) are:



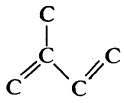
fulvene, C_6H_6 (bright yellow)



azulene, $C_{10}H_8$ (deep blue)

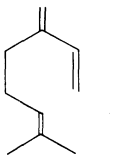
The colour depends both on the number of double bonds and on their arrangement in the molecule.

Occurrence and synthesis. A rather large number of hydrocarbons with more than one double bond occur in nature. Many of them belong to a class of compounds called isoprenoids. These substances have ten carbons or more (in multiples of five) with structures based on the linking together of isoprene units, fragments with the carbon atoms arranged as in isoprene (2-methyl-1,3-butadiene).

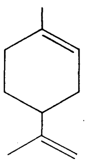


isoprene unit

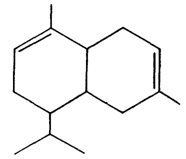
Examples of these structures, together with their common names and sources, are given below, showing only the skeleton.



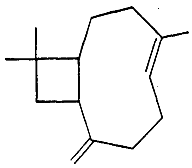
myrcene, $C_{10}H_{16}$ (bayberry)



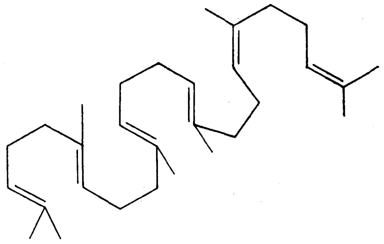
limonene, $C_{10}H_{16}$ (lemon, orange)



cadinene, $C_{15}H_{24}$ (oil of cade and of cubebs)



caryophyllene, $C_{15}H_{24}$ (oil of cloves)



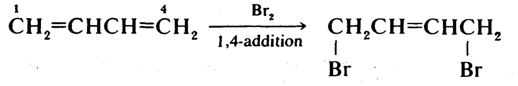
squalene, $C_{30}H_{50}$ (whale oil)

The most important commercial diene is 1,3-butadiene. It is manufactured by dehydrogenation of butenes:

$$\left.\begin{array}{l}CH_2{=}CHCH_2CH_3\\ \quad\text{and/or}\\ CH_3CH{=}CHCH_3\end{array}\right\}\xrightarrow{\text{calcium nickel phosphate, 650°}} CH_2{=}CHCH{=}CH_2 + H_2.$$

Isoprene is a product of hydrocarbon cracking, and it also can be obtained synthetically from propene, or from acetone and acetylene. Butadiene and, more recently, isoprene are used to manufacture synthetic rubber. Cyclopentadiene, a coproduct of butadiene and isoprene from cracking reactions, is the most readily available cyclic diene. Its chlorinated derivatives are used as insecticides (aldrin, dieldrin). Other dienes and polyenes can be synthesized by extensions of the reactions used to make simple alkenes.

Reactions. Dienes and polyenes undergo many of the reactions of simple alkenes. They add bromine, hydrogen, and acids; they can be converted to epoxides; and they undergo ozonolysis and polymerization. Conjugated dienes show certain special properties. In particular they show a propensity to add reagents to the ends of the conjugated system. The major product from the addition of one molecule of bromine to 1,3-butadiene, for example, is 1,4-dibromo-2-butene. The process is called 1,4-addition, in contrast to the usual addition to ordinary alkenes, which is called 1,2-addition. A double bond appears between carbons 2 and 3 in the addition product.

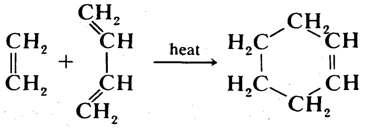


$$\overset{1}{C}H_2{=}CHCH{=}\overset{4}{C}H_2 \xrightarrow[\text{1,4-addition}]{Br_2} \underset{Br}{\underset{|}{C}}H_2CH{=}CH\underset{Br}{\underset{|}{C}}H_2$$

1,4-dibromo-2-butene

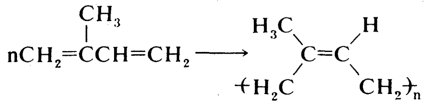
1,4-addition reactions

Conjugated dienes undergo 1,4-cycloaddition with alkenes. For example, 1,3-butadiene and ethylene give cyclohexene:



cyclohexene

The reaction is general and constitutes the most important method for making 6-membered rings. It was discovered by, and is named after, the German chemists Otto Diels and Kurt Alder (Diels–Alder reaction), who received a Nobel Prize in 1950. In the presence of certain catalysts, isoprene undergoes 1,4-polymerization to form a product indistinguishable from natural rubber, in which the isolated double bonds have the *cis* geometry.



$$nCH_2{=}\overset{CH_3}{\overset{|}{C}}CH{=}CH_2 \longrightarrow \left(\begin{array}{ccc} H_3C & & H \\ & C{=}C & \\ H_2C & & CH_2 \end{array}\right)_n$$

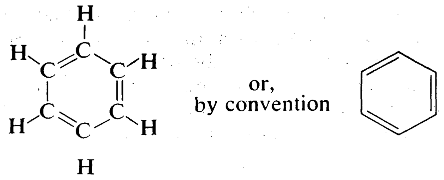
isoprene rubber

AROMATIC HYDROCARBONS

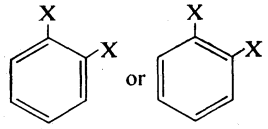
It was recognized during the first half of the 19th century that there exists a fairly large class of organic substances with distinctly different properties from the aliphatic compounds. The hydrogen–carbon ratio in the formulas of these compounds suggests unsaturation—it is often less than one, whereas in alkenes it is two, and it is even higher in alkanes. Nonetheless, these substances do not react as if they were unsaturated: they do not readily add bromine, are not easily oxidized, and mainly undergo substitution rather than addition reactions. Frequently these substances have pleasant aromas; for example, they can be obtained from the volatile oils of cloves, cinnamon, anise, wintergreen, vanilla, etc., and hence they were often called aromatic compounds. Joseph Loschmidt, an Austrian chemist, recognized in 1861 that most aromatic substances have formulas that can be derived from that of the hydrocarbon benzene, C_6H_6, by replacing one or more of the hydrogens of that substance by other atoms or groups. Following this observation the term aromatic lost its orginal, nontechnical meaning, and to the chemist came to mean any compound structurally derived from benzene. In the 1930s the concept of aromaticity was broadened considerably, and now both benzenoid and nonbenzenoid aromatic compounds are recognized. The former constitute the larger, and technically more important, class.

Origin of the term aromatic compounds

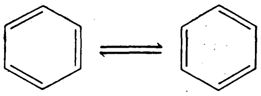
Benzenoid aromatic compounds. *Structure of benzene.* Friedrich August Kekule, the German chemist, suggested in 1865 that each carbon atom in benzene was at the corner of a regular hexagon, with one hydrogen atom attached to each carbon. To give each carbon a valence of four, Kekule inserted a conjugated system of double bonds into the ring. This formula satisfactorily explained the observation that benzene forms only one monosubstitution product (*e.g.,* one C_6H_5Cl, one C_6H_5Br, and one C_6H_5OH). All six hydrogens occupy equivalent po-



sitions in the Kekule formula. But it does not explain certain other observations; for example, benzene does not decolourize bromine rapidly at room temperature as other unsaturated molecules do. Nor does it explain why, when two adjacent hydrogens are replaced by other atoms (say X), only one compound is obtained. Kekule's formula predicts two isomers, depending on whether a double or a single bond separates the substituted carbons:

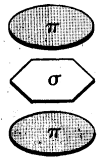


To overcome these difficulties Kekule suggested in 1872 that the positions of the double bonds were not fixed, but the two forms interconverted so rapidly that they could not be separated.



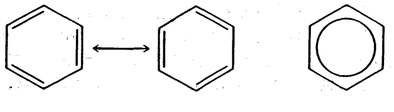
Kekule forms in equilibrium

The current view of the benzene structure differs from that of Kekule in an important conceptual way, though his formulas are still useful. Instead of two kinds of molecules in rapid equilibrium, benzene and its derivatives consist of only one kind of molecule, with a structure that can be considered a hybrid, called a resonance hybrid, of the two Kekule forms. All the carbon–carbon bonds in benzene are identical; they are neither single nor double, but of a new, hybrid type (superficially, one-and-a-half bonds). The most direct evidence for this structure comes from the molecular geometry. It has been shown that the carbon atoms in benzene are in fact at the corners of a regular hexagon, and the carbon–carbon bond distances are all identical at 1.39 Å, a value intermediate between those of the normal C—C and C═C distances (see above *Geometry of the double bond*). The hydrogen atoms of the molecule lie in the same plane as the carbons, and all angles (between carbon–hydrogens as well as between carbon–carbon bonds) are 120°. Looked at in another way, the six carbons in a benzene ring are joined by six σ bonds (six electron pairs that lie in the plane of the ring, one pair between each adjacent pair of carbon atoms); in addition, there are six π electrons (one pair for each of the three double bonds in a Kekule structure), and these lie in three π orbitals, occupying the space above and below the molecular plane. The π electrons are said to be delocalized, or spread over all six carbons.

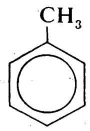


Space occupied by the
σ and π electrons in benzene

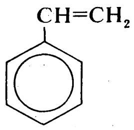
The hybrid structure of benzene is represented in various ways: one is to show a double-headed arrow between Kekule structures; another is a hexagon with a circle in the middle, to denote the delocalization of the π electrons. In the rest of this article, the latter structure is used to emphasize the delocalization of the double bonds.



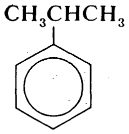
Nomenclature. Benzenoid aromatic hydrocarbons often are called arenes. Since this branch of organic chemistry developed at an early date, before aliphatic chemistry in general and before the development of systematic nomenclature, common names are used more frequently here than they are with aliphatic compounds. Some examples are:



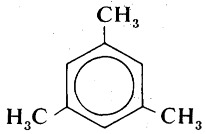
toluene
(methylbenzene)



styrene
(vinylbenzene)

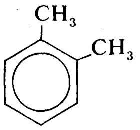


cumene
(isopropylbenzene)

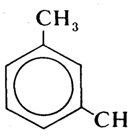


mesitylene
(1,3,5-trimethylbenzene)

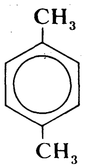
Two substituents may have three possible orientations on a benzene ring, called *ortho* (*o*), *meta* (*m*), or *para* (*p*), as illustrated with the three dimethylbenzenes, or xylenes:



o-xylene



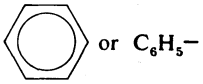
m-xylene



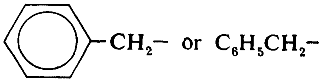
p-xylene

Two aromatic groups useful in naming compounds of this type are the phenyl and benzyl groups:

Phenyl and benzyl groups

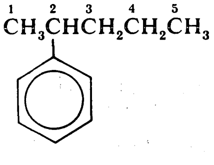


phenyl group

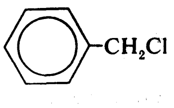


benzyl group

as in



2-phenylpentane



benzyl chloride

The benzene ring is sometimes called an aromatic nucleus, to differentiate it from a side chain or substituent that may be attached to it.

Physical properties. The smallest number of carbon atoms that an aromatic hydrocarbon can contain is six; hence none of these substances can have molecules small enough to be gases at room temperature (Table 15). Because of their symmetrical structures, benzene, *p*-xylene, durene, and hexamethylbenzene have relatively high melting points compared to those of closely related compounds that are more unsymmetrical. Toluene is used in place of mercury, which has a melting point of −38.9° C (−38° F), in low-temperature thermometers because of its low freezing point (−95° C [−139° F]) and its large coefficient of expansion (which means that large volume changes accompany small temperature changes). Aromatic hydrocarbons are better solvents for certain substances, such as lacquers, paints, and synthetic enamels, than are alkanes. Aromatic solvents need adequate ventilation, however, because even low concentrations are toxic, and prolonged exposure, especially to benzene vapours, can damage red blood cells.

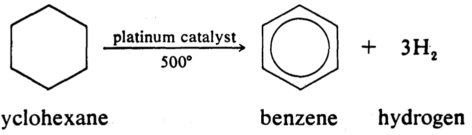
Table 15: Physical Constants of Benzene and the Methylbenzenes

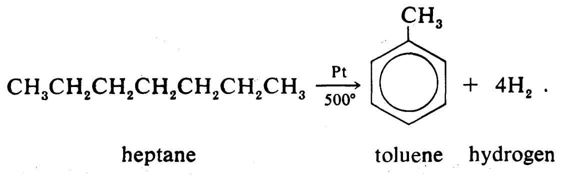
name	formula	boiling point (°C)	melting point (°C)
Benzene	C_6H_6	80.1	+5.5
Toluene	$C_6H_5CH_3$	110.6	−95
o-Xylene	1,2-$(CH_3)_2C_6H_4$	144.4	−25
m-Xylene	1,3-$(CH_3)_2C_6H_4$	139.1	−48
p-Xylene	1,4-$(CH_3)_2C_6H_4$	138.4	+13
Hemimellitene	1,2,3-$(CH_3)_3C_6H_3$	176.1	−25
Pseudocumene	1,2,4-$(CH_3)_3C_6H_3$	169.4	−44
Mesitylene	1,3,5-$(CH_3)_3C_6H_3$	164.7	−45
Prehnitene	1,2,3,4-$(CH_3)_4C_6H_2$	205	−6
Isodurene	1,2,3,5-$(CH_3)_4C_6H_2$	198	−24
Durene	1,2,4,5-$(CH_3)_4C_6H_2$	197	+80
Pentamethylbenzene	$C_6H(CH_3)_5$	232	+54
Hexamethylbenzene	$C_6(CH_3)_6$	265	+166

Source and synthesis. The chemistry of aromatic compounds was initially investigated during the last half of the 19th century, mainly in Germany and England, where illuminating gas and coke were produced by the large-scale carbonization of coal. Coal tar, a by-product of the process, contains about 25 percent by weight of aromatic hydrocarbons, together with other useful components.

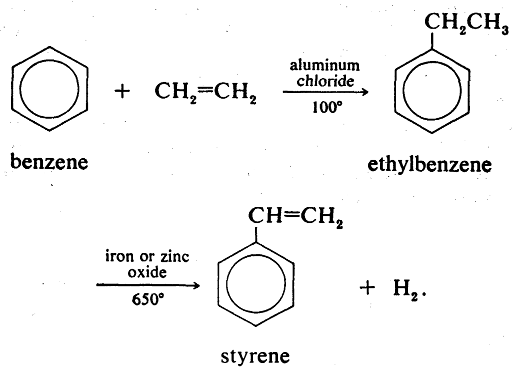
Aromatic hydrocarbons from petroleum

Currently, however, over 80 percent of aromatic hydrocarbon production is based on petroleum. Some aromatic hydrocarbons are present in crude petroleum and can be obtained directly, but most are produced by catalytic dehydrogenation of alkanes, a process called reforming:

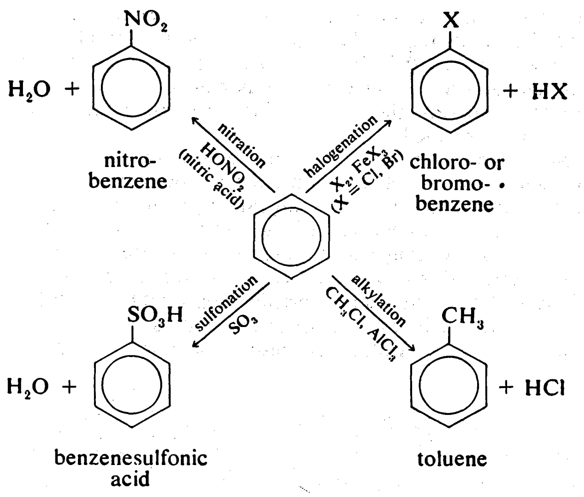




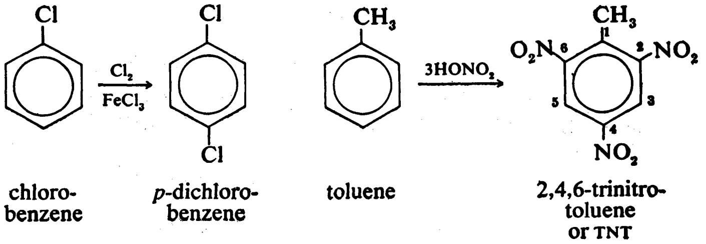
Other aromatic hydrocarbons can be obtained through synthesis, *e.g.*, ethylbenzene and styrene from benzene:



Reactions. The most characteristic reaction of aromatic hydrocarbons is substitution; *i.e.*, replacement of the aromatic hydrogens by other atoms or groups. Some of the more important substitution reactions are illustrated below:



The reaction may be repeated to replace additional hydrogens, producing di- or polysubstituted compounds, as in the chlorination of chlorobenzene and the nitration of toluene.

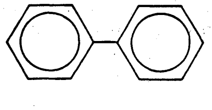


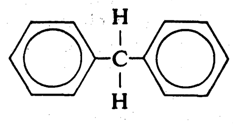
The products of these substitution reactions may be useful in their own right or as precursors of other useful materials, such as synthetic dyes, plastics, fabrics, and medicinals.

Under special conditions, aromatic hydrocarbons undergo addition reactions like those of unsaturated compounds. Thus benzene compounds can be catalytically hydrogenated to cyclohexanes, although higher temperatures and pressures are required than is the case with hydrogenation of simple alkenes. In the presence of light, benzene adds six atoms of halogens; the product from chlorine (1,2,3,4,5,6-hexachlorocyclohexane) is a widely used insecticide (Lindane).

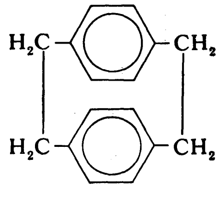
Polycyclic aromatic hydrocarbons. Many aromatic hydrocarbons with more than one benzene ring are known. Two or more rings may be joined to one another directly through a single carbon–carbon bond, as in biphenyl, or with one or more carbon atoms between the rings, as in diphenylmethane or paracyclophanes. More important commercially are the aromatic hydrocarbons with fused or condensed rings, in which two or more carbon atoms are shared in common by several aromatic rings. Naphthalene, anthracene, phenanthrene, dibenzanthracene, and

Fused ring compounds


biphenyl



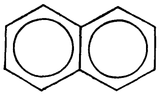
diphenylmethane

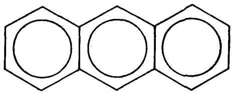


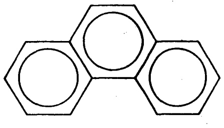
a paracyclophane

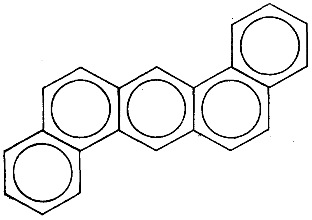
coronene are well-studied examples. As the number of fused rings increases, the carbon–hydrogen ratio also increases—from 1.0 in benzene to 2.0 in coronene. If the hexagonal network in coronene is extended indefinitely in all directions, the flat, cellular structure of graphitic carbon is obtained.

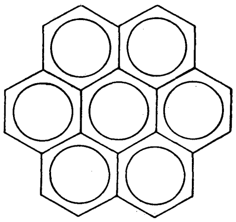
Most condensed aromatic hydrocarbons are beautifully crystalline solids. Many are present in coal tar; naphthalene is the most abundant, constituting about 10 percent of the tar, or 40 percent of the available aromatic hydrocarbons. Although most naphthalene is used to make polymeric resins and plasticizers employed in the manufacture of products made of plastics, appreciable quantities are consumed in the manufacture of dyes, insecticides, and medicinals. Anthracene, also present in coal tar, is colourless when pure; it exhibits a strong, pale-blue fluorescence when exposed to ultraviolet light. Many synthetic dyes contain the anthracene ring system. The phenanthrene ring system, especially in partially or fully hydrogenated form, is present in many natural products, such as the steroid hormones. Some fused polynuclear aromatic hydrocarbons, such as the dibenzanthracene shown, are carcinogenic—they produce skin cancers when applied locally to the skin and sarcomas (cancers of muscle, bone, or connective tissue) when injected subcutaneously in mice.


naphthalene
$C_{10}H_8$


anthracene
$C_{14}H_{10}$

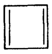

phenanthrene
$C_{14}H_{10}$

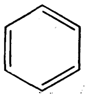

dibenzanthracene
$C_{22}H_{14}$

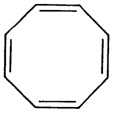

coronene
$C_{24}H_{12}$

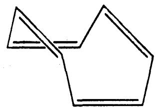
The same types of substitution reactions common with benzene also can be carried out with polynuclear compounds. Condensed ring hydrocarbons, such as naphthalene, usually react more readily than benzene does.

Nonbenzenoid aromatic compounds. Once it became clear that the aromatic properties of benzene and related compounds were due to the closed system of conjugated double bonds in the six-membered ring, it was only natural that organic chemists should attempt to synthesize both larger and smaller rings with conjugated systems of double bonds. The first targets were cyclobutadiene and cyclooctatetraene, the next lower and higher analogues of benzene. Cyclobutadiene has proved particularly elusive; most attempts to prepare it or its derivatives have failed. Cyclooctatetraene was first synthesized by Richard Willstätter, a German chemist, winner of a Nobel Prize in 1915. Although this substance is fairly stable, it does not have a planar conformation like that of benzene; rather, the molecule is tub-shaped. Cyclooctatetraene readily decolourizes bromine, undergoes addition rather than substitution reactions, and, in short, behaves like a polyene rather than an aromatic compound.


cyclobutadiene
C_4H_4

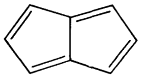

Kekule's formula of benzene
C_6H_6

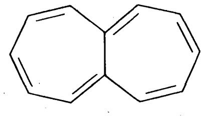

cyclooctatetraene
C_8H_8


cyclooctatetraene

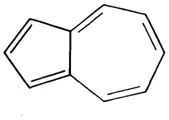
The Hückel rule

Some nonbenzenoid hydrocarbons with definitely aromatic properties are known, however. Erich Hückel, a German chemist, deduced theoretical reasons why cyclic, planar, conjugated systems must have $4n+2\pi$ electrons (*i.e.,* 2, 6, 10, 14 . . .) and not $4n$ π electrons (*i.e.,* 4, 8, 12 . . .), or any other number, if they are to be aromatic. Thus, cyclobutadiene, with four π electrons, and cyclooctatetraene, with eight, are not aromatic; whereas benzene, a planar molecule with six electrons, is aromatic. Other examples that conform to the Hückel rule are:


pentalene, C_8H_6
(8 π electrons, planar)

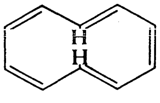

heptalene, $C_{12}H_{10}$
(12 π electrons, planar)

unstable, nonaromatic

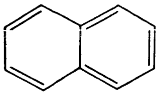

azulene $C_{10}H_8$
Kekule-type formula
(10 π electrons, planar)
stable, aromatic

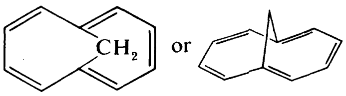
Thus, azulene, a nonbenzenoid compound, isomeric with naphthalene, readily undergoes substitution reactions, a property of aromatic rather than unsaturated systems.

Cyclodecapentaene, with ten π electrons, also should be aromatic according to the Hückel rule. But the two hydrogens that jut into the middle of the ring make this structure difficult to synthesize. If the two hydrogens are replaced with a bond, the resulting compound is naphthalene, which of course is aromatic, but benzenoid. On the other hand, if a single carbon atom, necessarily also bearing two hydrogens, is substituted for the two hydrogens of cyclodecapentaene, the conjugated ten-π-electron system can still be nearly planar. In striking confirmation of the Hückel theory, this hydrocarbon, though nonbenzenoid, undergoes substitution reactions and has other aromatic properties.



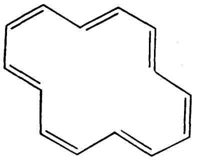
cyclodecapentaene

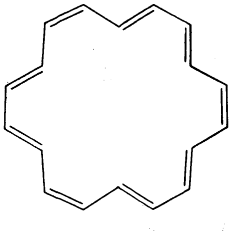

naphthalene



aromatic

Among the larger cyclic polyenes, called annulenes, the ones with 14 and 18 π electrons also give some evidence of being aromatic and, in accord with the Hückel rule, are more stable than those with 16 or 20 π electrons.


[14]-annulene
$C_{14}H_{14}$


[18]-annulene
$C_{18}H_{18}$

In general, it may be said of hydrocarbons that although their molecules are composed of atoms of only two elements, these atoms can be arranged in an almost infinite number of ways to give a class of organic compounds having an extraordinary variety of physical and chemical properties. (H.Ha.)

Alcohols, phenols, and ethers

The hydroxyl group

Alcohols, phenols, and ethers are organic chemical compounds—that differ from all other organic compounds in that their molecules contain at least one oxygen atom joined to a carbon atom solely by a single bond (as the connections between atoms are called). In alcohols and phenols this oxygen atom is part of a hydroxyl, or —OH, group; that is, the oxygen is joined to a hydrogen atom as well as to a carbon atom, $\equiv$C—OH. In ethers the oxygen atom is bonded to two carbon atoms, $\equiv$C—O—C$\equiv$.

Generally, the oxygen-bearing carbon atoms in alcohols, phenols, and ethers are part of hydrocarbon chains, series of carbon atoms to which hydrogen atoms are attached. Phenols differ from alcohols in that the hydrocarbons to which their oxygen atoms are joined are aromatic. Used in this sense, the word aromatic refers to certain rings of carbon atoms containing alternating single and double bonds, a system with distinctive properties.

It is instructive to compare the basic chemical structures of these compounds to the structure of the water molecule (see Table 16). Examination of these structures suggests that alcohols and phenols can be considered derivatives of water, in which a hydrocarbon chain has replaced a hydrogen atom. Experimentally, it is found that alcohols, in particular, resemble water closely, especially when the hydrocarbon group is small. In ethers, both hydrogen atoms have been replaced, and these substances do not resemble water in their behaviour at all.

Table 16: Comparison of the Structures of Water, Alcohols, Phenols, and Ethers

classification	structure*
Water	H—O—H
Alcohols	R—O—H
Phenols	Ar—O—H
Ethers	R—O—R′

*Ar = aromatic ring; R and R′ = hydrocarbon chains.

Another way of characterizing alcohols is to say that a hydroxyl (—OH) group has replaced a hydrogen atom in a hydrocarbon molecule (RH → ROH; the symbol R represents a group of atoms that remains unchanged in a reaction). The most typical alcohols are those in which a hydroxyl group is linked to a carbon atom attached to other atoms solely by single bonds. When the hydroxyl function is attached directly to an aromatic ring, as in phenols, the differences in physical and chemical properties are of such a magnitude that those compounds generally are not considered to be alcohols at all.

Alcohols are quite reactive substances chemically, and either the hydrogen of the hydroxyl group or the entire hydroxyl group can combine with or be replaced by other groups. Generally, alcohols are neutral compounds showing neither acid nor alkaline properties. Alcohols combine with inorganic compounds to form crystalline products, such as $MgCl_2 \cdot 6CH_3OH$, which are similar to common hydrate salts (inorganic compounds combined with water molecules). Primary alcohols are easily oxidized to aldehydes and to acids ($RCH_2OH \rightarrow RCHO \rightarrow RCO_2H$); secondary alcohols are oxidized to ketones ($R_2CHOH \rightarrow R_2C{=}O$); and tertiary alcohols on oxidation yield various decomposition products in which the carbon chain is broken ($R_3COH \rightarrow$ decomposition products). Alcohols generally react with carboxylic acids to produce neutral compounds called esters ($RCO_2H + R'OH \rightarrow RCO_2R' + H_2O$). Alcohols also can be converted to ethers under suitable dehydrating conditions ($2ROH \rightarrow ROR + H_2O$).

Every alcohol contains a grouping of atoms $\equiv$C—OH in which the three bonds on the carbon atom may be joined to three hydrogen atoms, three alkyl (hydrocarbon) groups, or any combination of these. Monohydric alcohols, those with single hydroxyl groups, can exist in various structural modifications. A primary alcohol contains at least two hydrogens on the carbon atom attached to the hydroxyl group; a secondary alcohol has one hydrogen on that carbon; there are no hydrogens on the hydroxyl carbon of a tertiary alcohol (see Table 17).

Table 17: Structure of Primary, Secondary, and Tertiary Alcohols

classification	examples	
	name	formula
Primary ($-CH_2OH$)	methyl alcohol	CH_3OH
Primary ($-CH_2OH$)	propyl alcohol	$CH_3CH_2CH_2OH$
Secondary ($>$CHOH)	isopropyl alcohol	$(CH_3)_2CHOH$
Tertiary ($\geq$COH)	trimethylcarbinol	$(CH_3)_3COH$

Acidic nature of phenols

Typical phenols have an acidic nature (as contrasted to alcohols and ethers, both of which are neutral, neither acidic nor basic). As acids, phenols exhibit a tendency to form salts with strong bases. Resonance structures (representations of the electronic structure of the molecule) of phenol itself and of the phenoxide ion, which is formed from phenol by loss of a proton, H^+, help explain the acidity of phenols. The resonance forms of phenol involving distribution of negative charge over the benzene ring can be produced only by a simultaneous development of positive charge on the oxygen atom. This means that separation of charges, an energy-consuming process, operates in producing these structures, and they make relatively little contribution to the overall state of the phenol molecule. With the phenoxide ion, on the other hand, no charge separation is involved; negative charge appears on all resonance structures, and those in which the charge is on the benzene ring serve to distribute charge already present. Thus, these forms make significant contributions to the structure of the ion, and it is favoured energetically over the phenol molecule. Since loss of the proton is what converts phenol to the phenoxide ion, the phenol molecule readily loses a proton or behaves in the fashion characterized as acidic.

The phenolic aromatic ring undergoes substitution reactions, such as nitration, chlorination, and sulfonation, more readily than benzene does, due to the presence of the negative charge on the ring, which serves to attract the substituting group. Because the negative charge on the ring appears only at certain locations (called the *ortho* and *para* positions: *ortho* for the ring carbon adjoining the one with the oxygen bond; *para* for the ring carbon diametrically opposite the one with the oxygen bond), it is primarily at these sites that substitution occurs. Under proper conditions, alkyl halides and olefins can react with phenols to form ethers. Esters of phenols are formed by reaction with acid chlorides or anhydrides.

Ethers are far less reactive than alcohols and phenols; they resemble hydrocarbons in that they are relatively inert at ordinary temperatures to strong bases, alkalies, and alkali metals. With strong acids, however, ethers form oxonium salts [$ROR' + HX \rightarrow (RR'OH)^+ + X^-$]. ($X^-$ stands for a negatively charged ion, such as chloride, Cl^-, or bromide, Br^-.) These react further with strong acids to produce alkyl halides and alcohols or water.

$$(RR'OH)^+X^- \xrightarrow{HX} RX + R'OH$$

$$(RR'OH)^+X^- + HX \xrightarrow{HX} RX + R'X + H_2O$$

ALCOHOLS

Nomenclature. The word alcohol is derived from the Arabic *kuḥl* (also *kohl* or *kohol*). It originally was used to mean a "very fine powder" but gradually came to connote "essence." Later the term was applied to wine spirits, which were referred to as *alcool vini,* and eventually simply alcohol.

A heterogeneous system of nomenclature has evolved for

Multiplicities of names

mono-, di-, and polyhydric alcohols, and the more important members have two or more common names. Ethanol, for example, is known as ethyl alcohol, grain alcohol, and spirits of wine. The first system of nomenclature for alcohols, which was based on the use of the term "carbinol" for methanol, was originated by a 19th-century German chemist, Hermann Kolbe. In this system, all alcohols are named as derivatives of carbinol, the various substituents being named first. Many secondary and tertiary alcohols are readily named by this system, as are aromatic and heterocyclic derivatives. Carbinol nomenclature is gradually fading out of usage, however.

The presently accepted systematic nomenclature for alcohols was adopted at a meeting of the International Union of Pure and Applied Chemistry (IUPAC), in Paris in 1957. According to the IUPAC system, the name of the alcohol is derived from the name of the parent hydrocarbon (which corresponds to the longest straight chain in the molecule incorporating the alcohol function). The final "-e" of the hydrocarbon name is replaced by "-ol" (or is followed by "-diol," "-triol," etc., for increasing numbers of alcohol functions). An arabic number before the name gives the position in the chain of the hydroxyl group, with the numbering beginning at the end of the chain that gives the hydroxyl the smallest number; *e.g.*, 2-butanol for $CH_3CH_2CHOHCH_3$. For unsaturated alcohols, the position of the hydroxyl group is given by a number before "-ol" and the number indicating the position of the double bond appears first, as 2-buten-1-ol for $CH_3CH{=}CHCH_2OH$. Certain other functional groups in the molecule, such as the carboxyl group, take precedence over the hydroxyl group in naming molecules and are given the terminal position; the prefix "hydroxy-" is used to denote the hydroxyl group. For example, 2-hydroxypropanoic acid is the systematic name for lactic acid, $CH_3CHOHCOOH$.

A frequent practice in forming common names is to indicate straight chains with a prefix "*n*-" (for normal) and certain branched chains with the prefix "iso-." The latter prefix is usually associated with those compounds containing a terminal $-CH(CH_3)_2$ group and no other branching. This prefix specifically applies to isopropyl, isobutyl, and isoamyl alcohol, but it is occasionally used in a more general sense to indicate any branched-chain structure.

Confusingly enough, the various alcohol nomenclature systems are in more or less simultaneous use. Generally, however, conversions from one form to another are made without difficulty (see Table 18).

Compounds containing two or more hydroxyl groups on separate carbons are named systematically according to the IUPAC rules as polyols, with separate numbers indicating the position of each hydroxyl. These compounds are also given common names, the diols being called glycols and the triols glycerols. Among the higher polyols are pentitols and hexitols.

Solubilities

General physical properties. Typical alcohols are colourless liquids or solids. Normal primary alcohols containing fewer than 12 carbon atoms are liquid at room temperature, whereas those with 12 or more carbons are solid at room temperature. Polyhydric alcohols usually are syrups; alcohols with complex arrangements of carbon atoms, such as the sterols (like cholesterol, which has 27 carbon atoms in its molecule), generally are solids. Lower alcohols, with up to three carbons in their molecules, are soluble in water, and all monohydroxy alcohols are soluble in organic solvents. The water solubility of alcohols decreases as the molecular weight rises, because the structure becomes relatively more hydrocarbon in nature. Conversely, the boiling points, vapour pressures, densities, and viscosities of alcohols increase with higher molecular weight. Adding hydroxyl groups to an alcohol without increasing the number of carbon atoms increases the solubility in water and decreases the solubility in ether and ethanol; at the same time the melting point, boiling point, density, and viscosity are increased. Hydrogen bonding—that is, secondary attractive forces between hydroxyl groups in adjacent molecules—is the main reason the alcohols have higher melting and boiling points than the corresponding hydrocarbons.

Additional hydroxyl groups in an alcohol enhance its sweetness. For example, ethanol, C_2H_5OH, is not sweet; propylene glycol, $C_3H_6(OH)_2$, is slightly sweet; glycerol, $C_3H_5(OH)_3$, is quite sweet; and mannitol, $C_6H_8(OH)_6$, is so sweet that it is known as a sugar alcohol. Lower alcohols have characteristic odours; alcohols with 8 to 12 carbons have roselike or lilylike odours, which find use in perfumes. The higher alcohols (with more than 12 carbons) are practically odourless. Physical properties of some common alcohols are shown in Table 19.

Occurrence in essential oils

Occurrence and preparation. Alcohols in free form are not a common occurrence in nature; they are found mainly in the essential or volatile oils obtained from the flowers, leaves, and stems of plants. Chiefly, these are a group of monohydric primary alcohols with carbon chains of 7 to 12 atoms. Their characteristic odours make them valuable ingredients for perfumes. More complex alcohols with branched structures and unsaturated linkages, such as citronellol and geraniol, both 10-carbon alcohols, are found in rose and geranium oils and are isolated from citronella oil. The fatty alcohols (primary alcohols with long, straight carbon chains), cinnamic alcohol ($C_6H_5CH{=}CHCH_2OH$), phenylpropyl alcohol ($C_6H_5CHOHC_2H_5$), and menthol and terpineol (cyclic 10-carbon alcohols) are other alcohols that occur naturally. Several sugar alcohols, containing, respectively, four and five hydroxyl groups on as many carbon atoms, are found in mosses, lichens, and yeasts. Sorbitol, a six-carbon, six-hydroxyl alcohol, occurs widely in fruits, berries, algae, and red seaweed. Mannitol, an isomer of sorbitol, is a major component of the exudate of plane, olive, and manna ash trees, and is found in pumpkin, grasses, mushrooms, marine algae, and brown seaweed. Sterols, such as cholesterol, are present in both the animal and vegetable kingdoms. They are found, for example, in bile, gallstones, brains, blood, egg yolk, and many plant products.

In nature, alcohols are many times more prevalent in combined form than as the free compounds. Fats, for instance, are esters of the trihydric alcohol, glycerol; and certain waxes and oils are esters of fatty alcohols. Many sterols occur in nature in esterified form.

Destructive distillation

Destructive distillation of hardwood was a common process in the early 1900s for producing wood alcohol as well as charcoal, acetic acid, and acetone. The major constituent of wood alcohol is methanol. It is quite difficult to purify this distillate to obtain the pure alcohol, however, so synthetic production has largely supplanted destructive distillation as the chief method used to make high-purity methanol in large quantities for industry.

Carbohydrates (sugar derivatives, such as starch) are a major group of natural compounds that can be made to yield ethyl alcohol by the biological process of fermentation. Before fermentation is begun, grain is cooked to form a starch mass; this mixture is cooled, and malt is then added to supply the enzymes that hydrolyze starch to fermentable sugar. Molasses also can be used as a starting material for fermentation. It contains enough sugar that

Table 18: Examples of Alcohol Nomenclature

formula	parent hydrocarbon	common name	carbinol name	IUPAC name*
CH_3OH	methane	methyl alcohol	carbinol	methanol
CH_3CH_2OH	ethane	ethyl alcohol	methylcarbinol	ethanol
$CH_3CHOHCH_3$	propane	isopropyl alcohol	dimethylcarbinol	2-propanol
$CH_3CH=CHCH_2OH$	2-butene	crotyl alcohol	propenylcarbinol	2-buten-1-ol
$(CH_3)_3COH$	2-methylpropane	*tert*-butyl alcohol	trimethylcarbinol	2-methyl-2-propanol

*Name approved by the International Union of Pure and Applied Chemistry.

Table 19: Physical Properties of Alcohols

IUPAC name*	common name	no. of carbon atoms	formula	melting point (°C)	boiling point (°C)	density (d_4^{20})
Methanol	methyl alcohol	1	CH_3OH	−93.9	64.96	0.7914
Ethanol	ethyl alcohol	2	CH_3CH_2OH	−117.3	78.5	0.7893
1-Propanol	*n*-propyl alcohol	3	$CH_3CH_2CH_2OH$	−126.5	97.4	0.8035
2-Propanol	isopropyl alcohol	3	$(CH_3)_2CHOH$	−89.5	82.4	0.7855
2-Propen-1-ol	allyl alcohol	3	$CH_2=CHCH_2OH$	−129	97	0.8540
1-Butanol	*n*-butyl alcohol	4	$CH_3(CH_2)_2CH_2OH$	−89.53	117.25	0.8098
1-Pentanol	*n*-amyl alcohol	5	$CH_3(CH_2)_3CH_2OH$	−79	138.0	0.8144
1-Dodecanol	lauryl alcohol	12	$CH_3(CH_2)_{10}CH_2OH$	26	259	0.8309^{24}
1-Octadecanol	stearyl alcohol	18	$CH_3(CH_2)_{16}CH_2OH$	59.4	332	0.8124^{59}
9-Octadecen-1-ol	oleyl alcohol	18	$CH_3(CH_2)_7CH=CH(CH_2)_7CH_2OH$	6–7	334	0.8489
1,2-Ethanediol	glycol	2	CH_2OHCH_2OH	−11.5	198	1.1088
1,2,3-Propanetriol	glycerol	3	$CH_2OHCHOHCH_2OH$	20	290	1.2613
Phenylmethanol	benzyl alcohol	7	$C_6H_5CH_2OH$	−15.3	205.35	1.0419

*Name approved by the International Union of Pure and Applied Chemistry.

it can be fermented without preliminary hydrolysis. In either case, yeast is added to bring about the fermentation; enzymes present in the yeast convert the sugars in the natural material to ethyl alcohol and carbon dioxide.

Fermentation

The fermentation process also may be used to produce other alcohols; *n*-butyl alcohol, for example, has been obtained from sterilized molasses by fermentation with a pure culture of bacteria (a species of *Clostridium*). Treatment of glycerol with certain other bacteria yields 1,3-propanediol, as well as *n*-butyl alcohol. Fusel oil, a mixture of impurities that occurs in fermentation ethanol, contains *n*-propyl, isobutyl, isoamyl, and active amyl alcohol (the *d*-isomer of this secondary alcohol). In Europe, glycerol has been produced as a by-product of the alcohol fermentation of sugar. The substance 2,3-butanediol is an example of a glycol obtained from sugar, molasses, or wheat mash by fermentation. Because of their occurrence in nature, many alcohols can be obtained from natural products by purely physical means such as distillation or, as indicated above, they can be liberated from complex chemical structures by means of enzymes, heat, or hydrolysis.

Synthetic production. Specific alcohols, such as ethanol, methanol, and glycerol, are consumed in enormous amounts for literally thousands of applications. Generally, the alcohols produced from natural sources are expensive, difficult to purify, and somewhat unpredictable in their availability. Therefore, in order to meet the demands of industry for alcohols of many descriptions in pure form and at low cost, producers have had to develop special processes that form alcohols from readily available raw materials under carefully controlled conditions.

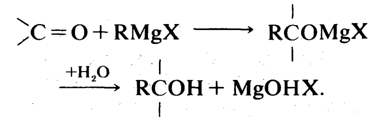
Aldol condensation. Aldehydes or ketones with a hydrogen atom on the carbon adjacent to the carbonyl group can be condensed to *β*-hydroxy aldehydes or ketones, respectively. Dehydration followed by catalytic hydrogenation then produces the alcohol; for example, acetaldehyde undergoes self-condensation to give, after dehydration, crotonaldehyde. This product then can be reduced to the alcohol 1-butanol.

$$\underset{\text{acetaldehyde}}{2CH_3CHO} \longrightarrow \underset{\text{condensation product}}{CH_3CHOHCH_2CHO}$$

$$\underset{\text{condensation product}}{CH_3CHOHCH_2CHO} \xrightarrow[-H_2O]{\text{acid}} \underset{\text{crotonaldehyde}}{CH_3CH=CHCHO}$$

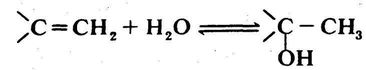
$$\xrightarrow[2H_2]{\text{reduction}} \underset{\text{1-butanol}}{CH_3(CH_2)_2CH_2OH}$$

The Grignard reaction. The Grignard reaction involves addition of an alkyl (or aryl) magnesium halide to an aldehyde, ketone, ester, or epoxide. Hydrolysis of the product produces an alcohol. In generalized terms, reaction of an alkylmagnesium halide with a carbonyl compound is as follows:



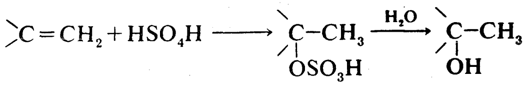
$$\rangle C=O + RMgX \longrightarrow R\overset{|}{\underset{|}{C}}OMgX \xrightarrow{+H_2O} R\overset{|}{\underset{|}{C}}OH + MgOHX.$$

Hydration of olefins (alkenes). A hydration process (addition of water to a double bond) is used industrially to produce ethanol directly from ethylene. The reaction is a general one and may be written as follows:



$$\rangle C=CH_2 + H_2O \rightleftharpoons \rangle\underset{OH}{C}-CH_3.$$

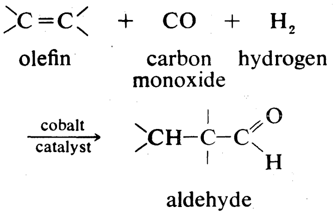
Indirect hydration may be carried out by adding the olefin to sulfuric acid, with subsequent hydrolysis of the intermediate alkyl sulfate.



$$\rangle C=CH_2 + HSO_4H \longrightarrow \rangle\underset{OSO_3H}{C}-CH_3 \xrightarrow{H_2O} \rangle\underset{OH}{C}-CH_3$$

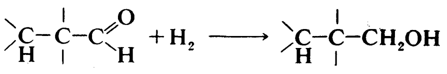
Secondary and tertiary alcohols may be produced in this manner.

Hydroformylation of olefins. Lower-molecular-weight olefins react with carbon monoxide and hydrogen in the presence of a catalyst in a reaction called hydroformylation or the oxo reaction.



$$\underset{\text{olefin}}{\rangle C=C\langle} + \underset{\text{carbon monoxide}}{CO} + \underset{\text{hydrogen}}{H_2} \xrightarrow[\text{catalyst}]{\text{cobalt}} \underset{\text{aldehyde}}{\rangle CH-\overset{|}{\underset{|}{C}}-C\genfrac{}{}{0pt}{}{\diagup\!\!\diagup O}{\diagdown H}}$$

The resulting aldehyde is subsequently hydrogenated to form an alcohol.



$$\rangle\underset{H}{C}-\overset{|}{\underset{|}{C}}-C\genfrac{}{}{0pt}{}{\diagup\!\!\diagup O}{\diagdown H} + H_2 \longrightarrow \rangle\underset{H}{C}-\overset{|}{\underset{|}{C}}-CH_2OH$$

Hydrolysis of halides and sulfates. Alkyl halides and sulfates readily undergo alkaline hydrolysis to yield alcohols.

$$RX + HOH \xrightarrow{\text{alkali}} ROH + HX$$

This general procedure produces primary and secondary alcohols. Glycerol can be synthesized from propylene by a series of reactions including hydrolysis of a halide as one step in the process. Unsaturated alcohols can be prepared by high-temperature chlorination followed by hydrolysis of the intermediate halide of an alkene. An example is the production of allyl alcohol from propylene.

$$CH_2{=}CHCH_3 + Cl_2 \longrightarrow CH_2{=}CHCH_2Cl + HCl$$

propylene chlorine allyl chloride hydrogen chloride

$$CH_2{=}CHCH_2Cl + NaOH$$

allyl chloride sodium hydroxide

$$\longrightarrow CH_2{=}CHCH_2OH + NaCl$$

allyl alcohol sodium chloride

Hydrolysis of organic esters. The hydrolysis of esters, which is catalyzed by acids and bases, at one time was of considerable importance in producing alcohols from natural esters. While not practiced commercially in the United States, it is employed in Europe and Japan.

$$R'COOR + HOH \longrightarrow R'COOH + ROH$$

Oxidation of hydrocarbons. Saturated hydrocarbons can be oxidized to alcohols in air by the use of high temperatures (260°–370° C [500°–698° F]), moderate pressures (15–50 atmospheres), and short periods of contact of the hydrocarbon with a catalyst.

Pyrolysis of fatty acids. In one particular instance the sodium salt of a hydroxyl-substituted, unsaturated fatty acid (ricinoleic acid) on pyrolysis gives an alcohol (2-octanol) among the products.

$$CH_3(CH_2)_5CHOHCH_2CH{=}CH(CH_2)_7COONa$$

$$\xrightarrow[\Delta]{NaOH} CH_3(CH_2)_5CHOHCH_3 + NaOOC(CH_2)_8COONa$$

2-octanol sodium sebacate

Reduction of carbonyl compounds. At high temperatures and pressures carbonyl compounds can be reduced in the presence of complex catalysts.

$$\rangle C{=}O + H_2 \longrightarrow \rangle CHOH$$

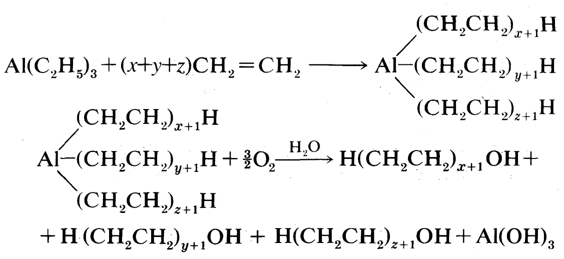
The largest application of this reaction is the production of methanol from carbon monoxide:

$$CO + 2H_2 \longrightarrow CH_3OH.$$

As carbonyl compounds, esters are reduced by alkali metals in alcohol or, in a more modern technique, by high-pressure hydrogenation over a copper oxide–chromium oxide catalyst. The process is used industrially to produce fatty alcohols from the esters of fatty acids.

$$RCOOR' + 2H_2 \longrightarrow RCH_2OH + R'OH$$

Catalytic polymerization of olefins. Since about 1955 straight-chain primary alcohols have been prepared in the petroleum industry from ethylene with the use of triethyl aluminum (Ziegler) catalysts. The course of the reaction is as follows:



$$Al(C_2H_5)_3 + (x+y+z)CH_2{=}CH_2 \longrightarrow Al\begin{cases}(CH_2CH_2)_{x+1}H\\(CH_2CH_2)_{y+1}H\\(CH_2CH_2)_{z+1}H\end{cases}$$

$$Al\begin{cases}(CH_2CH_2)_{x+1}H\\(CH_2CH_2)_{y+1}H\\(CH_2CH_2)_{z+1}H\end{cases} + \tfrac{3}{2}O_2 \xrightarrow{H_2O} H(CH_2CH_2)_{x+1}OH + H(CH_2CH_2)_{y+1}OH + H(CH_2CH_2)_{z+1}OH + Al(OH)_3$$

Principal chemical reactions. In general, the chemical reactions undergone by alcohols are essentially related to the hydroxyl groups and their position in the molecule. The nature of the hydrocarbon radical and the presence of other functional groups, however, also exert considerable influence. Alcohols are neutral compounds that do not ionize in water. Nevertheless, when they react with alkali metals, alcohols do act as acids, and when they react with strong acids they are functioning as bases. Tertiary alcohols are the most basic of the alcohols because the alkyl groups donate electrons to the carbon to which the oxygen is attached; this process increases the electron density at the oxygen atom and makes it more attractive toward a proton (*i.e.*, more basic).

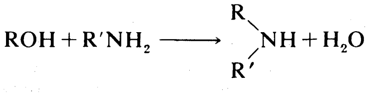
Neutrality of alcohols

Oxidation and the formation of esters and ethers (see above) are probably the most common reactions of alcohols. A number of other reactions also are encountered frequently in work with alcohols.

Ammonolysis and aminolysis. Alcohols generally react with ammonia to give amines and with amines to give further alkylation products.

$$ROH + NH_3 \longrightarrow RNH_2 + H_2O$$

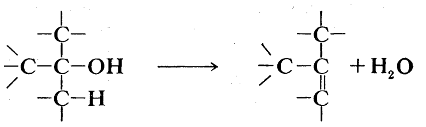
alkylamine



$$ROH + R'NH_2 \longrightarrow R(R')NH + H_2O$$

dialkylamine

Dehydration. Dehydrating agents remove water from alcohols with the production of unsaturated compounds or, if two molecules of alcohols are involved in the reaction, with formation of ethers. Tertiary alcohols tend to be dehydrated to alkenes the most easily; primary alcohols the least. For this reason, primary alcohols are converted to ethers more readily (without competition from alkene formation).



tertiary alcohol alkene

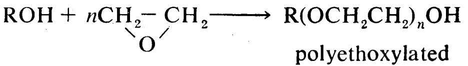
$$2ROH \longrightarrow ROR + H_2O$$

ether

Dehydrogenation. Like oxidation, dehydrogenation of primary alcohols gives aldehydes and dehydrogenation of secondary alcohols gives ketones. The high-temperature dehydrogenation of ethyl alcohol to acetaldehyde is a good example.

$$CH_3CH_2OH \xrightarrow{300° C} CH_3CHO + H_2$$

Ethoxylation. The reaction of alcohols with ethylene oxide gives polymeric products in which many units of the ethoxy group ($-CH_2CH_2O-$) are incorporated.



$$ROH + nCH_2{-}CH_2(O) \longrightarrow R(OCH_2CH_2)_nOH$$

polyethoxylated alcohol

Halogenation. Conversion of alcohols to alkyl halides can be accomplished by treatment with hydrogen halides or phosphorus trihalides.

$$ROH + HCl \longrightarrow RCl + H_2O$$

alkyl chloride water

$$3ROH + PCl_3 \longrightarrow 3RCl + P(OH)_3$$

Reaction with alkali metals. Alkali metals convert alcohols to their alkali metal alkoxides. Reaction of an alcohol with metallic sodium, for instance, produces the corresponding sodium alkoxide.

$$2ROH + 2Na \longrightarrow \underset{\text{sodium alkoxide}}{2RONa} + H_2$$

Sulfation. Treatment of alcohols with fuming sulfuric acid gives alkyl sulfates, a reaction analogous to ester formation with carboxylic acids.

$$ROH + H_2SO_4 + SO_3 \longrightarrow ROSO_3H + H_2SO_4$$

Toxicity

Biological and medical properties. With the exception of methanol, most monohydric saturated alcohols in industrial use are not hazardous as far as cumulative injury to internal organs is concerned, but some have a cumulative effect on the nervous system. The toxic effect of aliphatic alcohols on the central nervous system is a narcosis, generally increasing with the molecular weight of the alcohol. Apparently the increased effect with increasing molecular weight results from the larger hydrocarbon portion of those molecules, which favours solubility in lipids (fatty materials) of the nervous system. Physiologically, all alcohols are weaker narcotics than their parent hydrocarbons, but metabolic differences influence the degree and nature of toxicity. Ethyl alcohol, for example, is a noncumulative poison because the human body rapidly converts it to carbon dioxide and water. Methyl alcohol, by contrast, is labelled a poison because it has a slower rate of elimination and decomposes to formaldehyde and formic acid, both of which are poisonous to nervous tissue and are able to damage ocular tissues.

As the number of carbon atoms in an alcohol exceeds eight, the resultant low volatility and poor water solubility of the molecule reduces the likelihood of its ingestion or inhalation. There is a decrease of toxicity in a series of alcohols of the same number of carbon atoms in going from primary to secondary to tertiary. Unsaturation brings about a pronounced increase in toxicity, with allyl alcohol being about 30 times more toxic than *n*-propyl alcohol.

Alcohols as chemical intermediates

Industrial applications. As a group, alcohols are among the most important organic chemicals, judged by the large amounts produced. The most important industrial use of alcohols is as chemical intermediates, chiefly because alcohols can readily be converted into a large variety of other compounds by simple chemical reactions. For instance, reaction with carboxylic acids leads to fats and waxes; dehydration produces olefins and ethers; sulfonation and ethoxylation form surface active agents for use in detergents; and other syntheses yield compounds used as plasticizers, emulsifiers, lubricants, emollients, and foaming agents.

Ethyl alcohol is second in importance only to water as a solvent for drugs, lacquers, perfumes, cosmetics, detergents, and plastics. Recent applications are in aerosol and mouth-wash products, a motor fuel, and a rocket fuel.

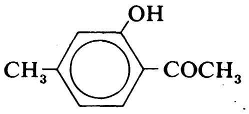
Identification and analysis. Physical methods, such as spectroscopy and chromatography, are frequently used to identify alcohols. Chemically this is done by forming a derivative that characterizes the functional hydroxyl group. An important first step is to convert the crude alcohols or more complex mixtures containing other substances to organic or inorganic esters, which can then be isolated in pure form. Additionally, if the esters are crystalline in nature with well-defined melting points, the melting points may serve for purposes of identification. Esters frequently used for the characterization of alcohols are the phthalates, benzoates, and borates. Carbamates, commonly known as urethans, are popular derivatives of primary and secondary alcohols, particularly *α*-naphthylurethan and *p*-nitrophenylurethan. Formation of complex crystalline salts with calcium chloride is another technique for isolation and purification of alcohols.

Detection of impurities

Often in analysis of an alcohol the substances of interest are the impurities rather than the alcohol itself. Hence, many analytical methods involve separation of minor components by laboratory manipulations, whereas the amount of alcohol is determined by a simple reading on an instrument. The most significant and useful determination involves the hydroxyl function; it is analyzed most frequently by acetylation, phthalation, formylation, and oxidation reactions or by the measurement of active hydrogen. Colour, odour, taste, refractive index, density, and melting and boiling points are properties that aid in identifying alcohols and their states of purity.

PHENOLS

Nomenclature. Phenol is both the specific name for monohydroxybenzene (C_6H_5OH) and the generic name for compounds containing one or more hydroxyl functions attached to an aromatic ring. As with alcohols, nomenclature has been built around trivial names, consequently such names as phenol, cresol (for methylphenols), and naphthol (for hydroxynaphthalenes) are widely used and have been adopted as part of the IUPAC rules. In systematic nomenclature for phenols with more elaborate ring systems, the name for the phenol is obtained by dropping the final "e" from the hydrocarbon name and adding "ol"; thus, the phenols formed from chrysene are termed chrysenols. Frequently, in naming phenols with many substituents, the hydroxyl group is treated as merely another substituent; in this case it is denoted by the prefix "hydroxy-" and a number to locate it on the ring as 4-methyl-2-hydroxyacetophenone for the compound.



General physical properties. Phenols are colourless liquids or white solids at room temperature. Monohydric phenols have a characteristic burning taste. The lowest member of the family, phenol, has a bland and sweetish smell, whereas higher homologues are sharp and spicy smelling. Dihydric phenols, such as resorcinol and hydroquinone, have only slight odours; and trihydric and other polyhydric phenols have essentially no odour at all. Phenols have considerably higher melting points, boiling points, and densities than the parent hydrocarbons from which they are derived (see Table 20).

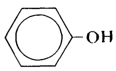
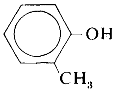
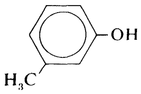
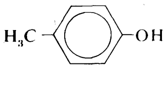
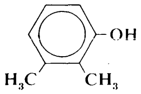
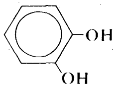
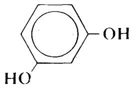
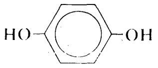
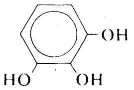
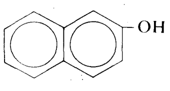
Since the phenolic hydroxyl group is more strongly polarized than that of alcohols, phenols tend to form stronger hydrogen bonds than do alcohols. As a result, phenols show increased solubilities in water and higher boiling points than alcohols, and their solvency for polar organic molecules is greater. Monohydric phenols are partially soluble in water, and polyhydric phenols are much more soluble. Most phenols are readily soluble in oxygen-containing organic solvents but less so in nonpolar solvents such as the aliphatic hydrocarbons. Alkyl substitution on the benzene ring of phenols, however, enhances their solubility in nonpolar solvents.

Phenols have antiseptic and germicidal properties, which increase to a maximum as the length of an alkyl side-chain substituent reaches about six carbon atoms. The lowest member, phenol, is highly toxic and caustic, but these properties progressively diminish with the higher members of the series. Polyhydric phenols are still markedly toxic but less caustic than the monohydric compounds.

Sources of "natural" phenol

Occurrence and preparation. So-called "natural" phenol is obtained by the destructive distillation of coal or from wood tar although it also can be obtained from petroleum and from organic matter of vegetable origin. Other phenols of truly natural origin (that is, not dependent on prior chemical transformation) are found in essential oils, particularly thymol and carvacrol. Phenol itself was first obtained from coal tar in 1834 by the German chemist Friedlieb Ferdinand Runge and called carbolic acid. The French chemist Auguste Laurent in 1841 produced the pure compound from coal tar and named it phenic acid. Although coal tar represented the major source of phenol before World War I, the great use of phenol in plastics spurred the use of synthetic methods of production, with the result that by the mid-1960s over 95 percent of the phenol produced was of synthetic origin. Three general procedures are used for synthesis, each of which utilizes benzene as a starting material; the methods are: sulfonation, chlorination, and alkylation with propylene (which leads to cumene, or isopropylbenzene, as the first intermediate).

Table 20: Physical Properties of Phenols

	formula	melting point (°C)	boiling point (°C)	density (d^{20}_4)
Phenol		43	181.75	1.0576
o-Cresol		30.94	190.95	1.0273
m-Cresol		11.5	202.2	1.0336
p-Cresol		34.8	201.9	1.0178
2,3-Dimethyl-phenol		75	218	1.16
Pyrocatechol		105	245*	1.1493 (21°)
Resorcinol		111	280	1.2717
Hydroqui-none		173–174	285–287	1.328 (15°)
Pyrogallol		133	309	1.453 (4°)
2-Naphthol		123	295	1.28

*At 750 mm Hg.

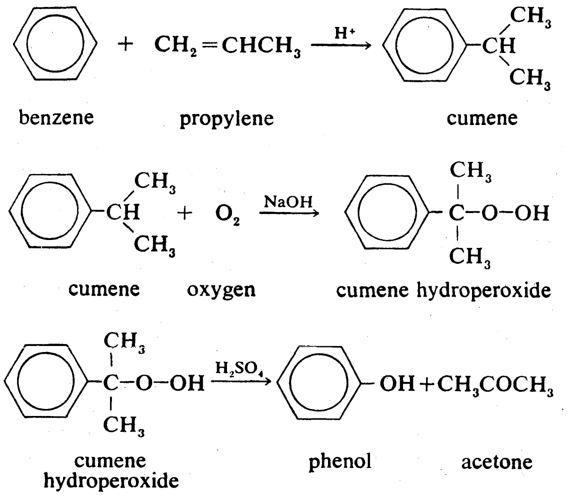
In addition to phenol itself, the cresols, xylenols, and higher phenol homologues are obtained from coal tar and other products of the destructive distillation of organic matter (usually of vegetable origin). Coal tar has been a prime source of phenols because of its extensive production in gasworks and coke-oven plants, which heat coal to produce heating gases for fuel and coke for steel manufacture. Depending on the source of the coal and the conditions used for carbonization, various complex mixtures of acidic tars are obtained along with tar bases and naphthalene. When this coal tar is distilled, a "middle oil" fraction is obtained that is rich in phenols. Further treatment with aqueous sodium hydroxide and acidification with carbon dioxide separate phenol from neutral and basic impurities and subsequent distillation yields phenol, cresols, and cresylic acids.

Phenols in free form occur in the so-called essential oils, which are derived from seeds or leaves of plants. Oils of thyme and origanum may contain from 60 percent to 80 percent of thymol or carvacrol. Thymol is also obtained in substantial quantities from Indian ajowan seed oil. Modern demand for these natural phenols has spurred synthetic processes, which are used to supplement material from natural sources. A nonvegetable source of phenols is petroleum distillates; several processes exist for isolating phenols from petroleum tars.

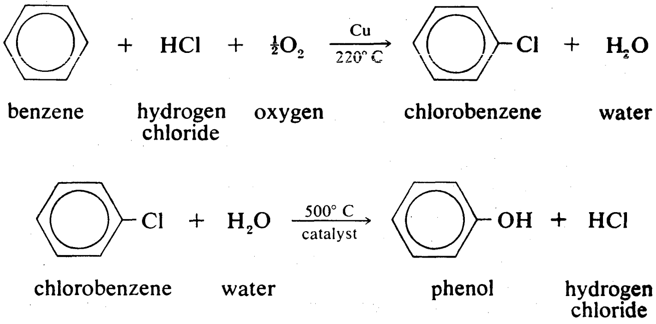
Synthetic production. The following synthetic methods are widely used to make phenol and, on a far smaller scale, certain other monohydroxybenzene derivatives.

Oxidation of cumene

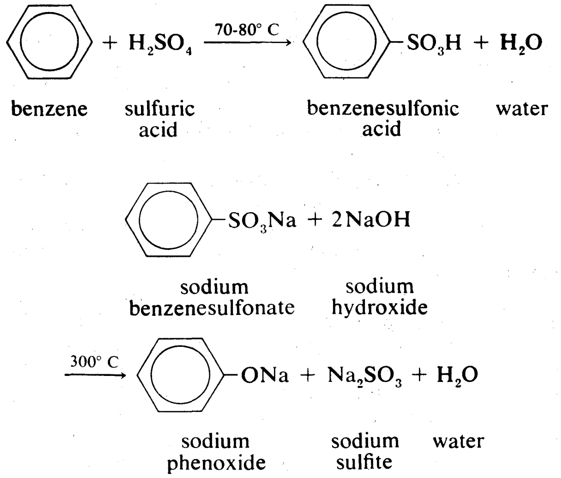
The cumene process. In the first step in this procedure, benzene is alkylated with propylene to give cumene. Cumene is then oxidized to cumene hydroperoxide by air in the presence of an alkaline catalyst, and the hydroperoxide is cleaved with an acid catalyst to phenol and acetone. This process may also be used to make cresols from toluene.



Chlorination. In a modification of the old German Raschig process, chlorination of benzene is accomplished with hydrogen chloride and oxygen. Chlorobenzene is hydrolyzed in the vapour phase over a catalyst to give phenol and regenerate hydrogen chloride.



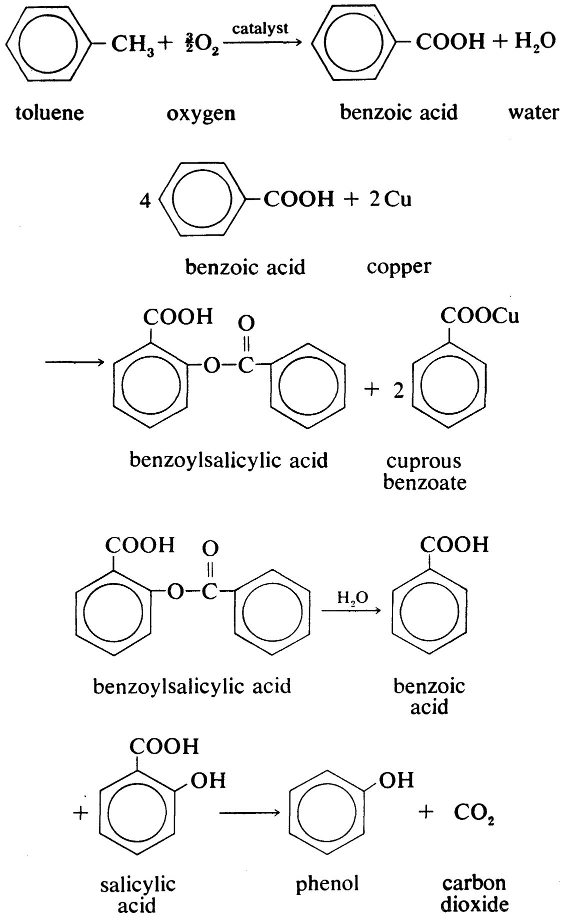
In an alternate process the intermediate chlorobenzene is purified by distillation and hydrolyzed with aqueous sodium hydroxide at high temperature and pressure. Formation of a by-product, diphenyl ether, is repressed by recycling this compound in the feed to maintain an equilibrium concentration. Sodium phenoxide is separated by



solvent extraction and acidified with hydrochloric acid to give phenol, which is purified by distillation.

Sulfonation. Benzene is sulfonated in the liquid state with sulfuric acid, and then neutralized with sodium sulfite. The resulting sodium benzenesulfonate is fused with sodium hydroxide, and the phenoxide neutralized with sulfur dioxide to phenol.

Toluene oxidation. Toluene is oxidized to benzoic acid by the use of air at 130–140° C (266–284° F) with a cobalt salt catalyst. The second step employs air, steam, and a copper–magnesium salt mixture to form an intermediate cupric benzoate, which decomposes to cuprous benzoate and benzoylsalicylic acid. Steam hydrolyzes the acid ester to benzoic and salicylic acids, and the latter compound decomposes to phenol and carbon dioxide:



Cyclohexane oxidation. Cyclohexane is oxidized with air at 130° C in the liquid phase with a cobalt naphthenate catalyst to cyclohexanol and cyclohexanone. The crude product is purified by distillation and dehydrogenated at 380° C (716° F) over a platinum catalyst. The reaction mixture is fractionated to yield phenol.

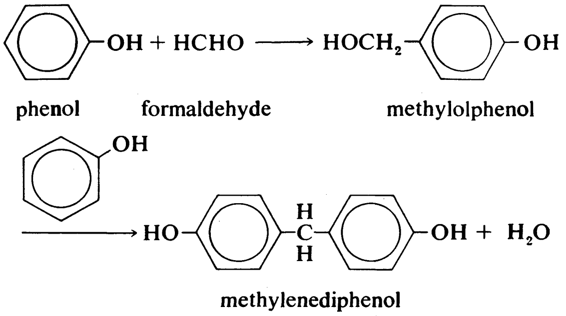
Principal chemical reactions. The chemical properties of phenols differ from those of alcohols chiefly because of the acidic character of the phenols. Because of their acidity, the phenols react with strong bases to form alkali-metal salts known as phenoxides, phenolates, or phenates. These salts are readily decomposed by acids or carbon dioxide.

Phenols alkylate more readily than their parent hydrocarbons, primarily in the *ortho* and *para* positions. Alkylation with dienes forms alkenyl phenols or cyclic compounds or both. Alkyl halides or dialkyl sulfates form phenolic ethers readily by reaction with alkali metal phenoxides. Phenol esters result from reactions of phenols with acid chlorides or anhydrides.

As a simple derivative of benzene, phenol can be readily chlorinated, nitrated, and sulfonated. It can be oxidized to give di- and trihydric phenols, derivatives of diphenols, and various decomposition products. Catalytic hydrogenation produces cyclohexanol and cyclohexane.

Phenol–formaldehyde resins

The most important reaction of phenol is its condensation with formaldehyde to form useful synthetic polymers, the phenol–formaldehyde resins, which are widely used to make articles of plastic. Under alkaline conditions, phenol reacts with formaldehyde to form methylolphenols and methylenediphenol.



Antiseptic properties

Biological and medical properties. The antiseptic activity of phenol is a property that accounts for much of its industrial utilization. In dilute solutions (of 1 to 2 percent) it also finds application as an agent for combatting itching of the skin. Undiluted phenol is highly corrosive to mucous membranes and skin, and it is considered a nerve poison. It may enter the human system by many methods, including absorption by skin, oral ingestion, and vapour inhalation. Phenol and the three cresol isomers are of about the same order of toxicity and produce identical symptoms in poisoned animals.

The lethal dose of phenol for adults has been estimated at 10 to 15 grams (0.32 to 0.48 ounces) orally and less when absorbed from body cavities or mucous membranes. Systemic toxicity of ingested phenol depends on the absolute amount consumed. Continuous exposure even to low levels may cause a chronic poisoning. A safe allowed concentration of phenol vapour in air for an eight-hour exposure is 5 parts per 1,000,000. The metabolic effects of phenol are not well known.

Industrial applications. In the 19th century, phenol was used extensively as an antiseptic, and the term "phenol coefficient" is widely used as a measure of the relative strength of disinfectants. Due to its irritating effect on humans, the use of phenol has been largely discontinued for general antisepsis. Higher-molecular-weight homologues and derivatives of phenol have supplanted phenol itself in the areas of germicidal cleaning.

World War I created a large demand for phenol as a starting material for production of a trinitro derivative, picric acid, which was used extensively as an explosive. In the 1930s the expanding plastics industry began to require large amounts of phenol for the Bakelite-type resins, and this usage expanded to the point in the 1950s where it consumed about 70 percent of total phenol production. From the 1960s onward this use fell significantly.

The chemical and pharmaceutical industries are large users of phenol for conversion to many different products. The salicylates (carboxyphenols) are starting materials for the preparation of aspirin, dyes, and flavours. Chlorophenols and their derivatives find application as fungicides, bactericides, and selective weed killers. Alkyl phenols produce an important group of synthetic detergents and are also used as oxidation inhibitors. Phenolsulfonic acid can be converted to synthetic tanning agents, and triphenyl phosphate is a plasticizer for cellulose acetate film. Large quantities of phenol are converted to cyclohexanol, which is converted eventually to nylon, and to bisphenol-A, used in the production of epoxy resins.

Identification and analysis. Natural phenol may contain small amounts of moisture, oils, tars, and naphthalene and substantial amounts of *o*-cresol; it normally assays in the range of 82 to 99 percent. Synthetic phenol, on the other hand, runs above 98 percent, with better grades assaying over 99.5 percent. The moisture content of phenol samples can be determined with a special reagent known as the Karl Fischer reagent. When a known impurity is present in phenol, depression of the freezing point provides a method of determining the amount.

Purification by means of acidity

Since phenols are weak acids, treatment with aqueous sodium hydroxide forms water-soluble phenolates, which may then be readily separated from water-insoluble compounds. Acidification regenerates phenol from the phenolate. This is a useful purification method.

Characteristic solid products result from the reaction of phenols with picric acid. The well-defined melting points of these products permit identification of the specific phenols that produce them. In the absence of other phenol compounds, a substitution reaction that yields the 2,4,6-tribromophenol can be used to determine the phenol on the basis of the amount of bromine consumed in the reaction.

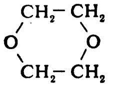
Qualitative determination of phenols is readily achieved by the well-known reaction with ferric chloride, which produces highly coloured compounds with the different phenols; phenol itself and resorcinol give violet colours, cresol forms a blue colour, and catechol yields a green-coloured complex. Instrumental methods, especially gas and liquid chromatography and infrared spectrophotometry, are widely used in the detection, identification, and analysis of phenols.

ETHERS

Nomenclature. The name ether apparently was applied to diethyl ether for the first time in 1730 by F.G. Frobenius, who referred to the compound as *spiritus aethereus,* or *vini vitriolatus.* The word ether is derived from the Greek *aithēr* meaning the clear sky or air; it represents the airy or volatile nature of typical compounds of this class. As with the term phenol, ether is used both for a specific compound (diethyl ether) and as a general term; in this case, however, use of the general term in the specific sense has not been officially sanctioned and is discouraged. In common nomenclature of ethers, simple symmetrical compounds with both R groups the same are named for the parent hydrocarbon; thus, $CH_3CH_2OCH_2CH_3$ is called diethyl ether or, simply, ethyl ether. With complex ethers containing different R groups both are named, as methyl ethyl ether for $CH_3OCH_2CH_3$. Alternatively, ethers may be designated by the IUPAC system as hydrocarbon derivations; $CH_3OCH_2CH_2CH_2CH_3$, for example, is called in this system 1-methoxybutane. Trivial or unsystematic names also are used for some complex ethers, such as anisole for methyl phenyl ether ($CH_3OC_6H_5$) and phenetole for ethyl phenyl ether ($CH_3CH_2OC_6H_5$).

General physical properties. At room temperature, ethers are colourless, neutral liquids with pleasant odours. Saturated and unsaturated aliphatic ethers have densities that range from 0.7 to 0.8, whereas the densities of aromatic ethers range from 1.0 to 1.1. Compared to the corresponding alcohols, ethers are generally lower in density, less water soluble, and have lower boiling points. The boiling points of ethers tend to approximate those for hydrocarbons of the same molecular weight, from which it can be concluded that association in liquid form is not
Solubilities significant. Ethers are easily soluble in organic liquids. Dioxane, however, is unusual among ethers in that it is completely soluble in water as well. A large number of organic compounds form azeotropes (constant-distilling mixtures) with ethers. Because of these properties, ethers are extensively used for the extraction and separation of organic chemicals. Physical properties of some common ethers are shown in Table 21.

Table 21: Physical Properties of Ethers

	formula	melting point (°C)	boiling point (°C)	density (d_4^{20})
Methyl ether	$(CH_3)_2O$	−138.5	−23.0	—
Ethyl ether	$(C_2H_5)_2O$	−116.2	34.51	0.7138
Isopropyl ether	$[(CH_3)_2CH]_2O$	−85.89	68.0	0.7241
n-Butyl ether	$(n\text{-}C_4H_9)_2O$	−95.3	142.0	0.7689
Vinyl ether	$(CH_2{=}CH)_2O$	−101	28	0.773
Allyl ether	$(CH_2{=}CHCH_2)_2O$	—	94.0	0.8260
Cyclopropyl methyl ether	$\overline{CH_2CH_2C}HOCH_3$	−119	44.7	0.8100
Benzyl ether	$(C_6H_5CH_2)_2O$	3.60	298	1.0428
Diphenyl ether	$(C_6H_5)_2O$	26.84	257.93	1.0748
1,4-Dioxane		11.8	101.1	1.0353

Occurrence and synthetic production. Simple ethers are not found commonly in nature, but the ether linkage is present in such natural products as sugars, starches, and cellulose.

Dehydration. Catalytic dehydration of alcohols is the oldest and most important method of producing ethers. Sulfuric acid is the dehydration catalyst most commonly used; it remains behind when water and ether are removed by distillation.

$$ROH + H_2SO_4 \xrightarrow{\text{cold}} RHSO_4 + H_2O$$

$$RHSO_4 + ROH \xrightarrow{\text{heat}} ROR + H_2SO_4$$

Williamson ether synthesis. The reaction of alkyl halides with sodium alcoholates is generally known as the Williamson ether synthesis. Both simple and mixed ethers are made in this manner. Metallic oxides may be used in place of the alcoholate to produce symmetrical ethers.

Reaction of olefins with alcohols or phenols. In the presence of sulfuric acid, alcohols or phenols combine with olefins to produce mixed ethers.

$$(CH_3)_2C{=}CH_2 + ROH \xrightarrow{H_2SO_4} (CH_3)_3COR$$

Reaction of boron trifluoride alcoholates. Heating of boron trifluoride alcoholates under pressure yields simple ethers.

$$n\text{-}C_4H_9OH \cdot BF_3 \xrightarrow[\text{autoclave}]{\text{benzene}} \underset{\substack{\text{dibutyl}\\ \text{ether}}}{(n\text{-}C_4H_9)_2O}$$

Use of diazomethane. Methyl ethers can be made from alcohols by reaction with the methylating agent, diazomethane.

Principal chemical reactions. Ethers are relatively inert with regard to chemical reaction, and in this regard they resemble the corresponding alkanes (which carry no functional groups). Since neither alkali metals nor alkalies attack ethers at low temperatures, they may be dried over metallic sodium.

Autoxidation. In the presence of oxygen, ethers oxidize to unstable peroxides, and this reaction may create the danger of explosion in stored ethers.

Salt formation. Strong inorganic acids and Lewis acids (compounds capable of accepting electron pairs) form oxonium compounds with ethers.

Cleavage reactions. Heat decomposes ethers, especially in the presence of alumina catalysts, to form olefins and water as the principal products.

Strong acids, such as hydriodic and nitric acid, cleave one or both of the carbon–oxygen linkages to form alkyl derivatives, and, in those cases in which only one bond is cleaved, alcohols.

Complex formation. Ether is used extensively as a solvent for Grignard reagents because it forms soluble complexes with most alkylmagnesium halides. It also forms complexes with inorganic halides, boron trifluoride, lithium aluminum hydride, and lithium borohydride.

Use as an anesthetic

Biological and medical properties. Ether has a widespread use as a surgical anesthetic; it acts on the central nervous system to render the patient insensible. Administration is by inhalation, rapid absorption taking place from the lungs. Ether is not a severe skin irritant, but it may cause slight irritation when swallowed or applied rectally. Ether is rapidly excreted from the lungs; small amounts also are excreted in the urine. The compound undergoes no chemical change in the human body. Concentrations of ethyl ether of more than 400 parts per 1,000,000 in air are considered dangerous when applied by inhalation. Ingestion of more than a fluid ounce may cause death.

Isopropyl and vinyl ethers are less effective anesthetics than ethyl ether, and the isopropyl compound is 1.5 to 2.0 times more toxic than ethyl ether. *n*-Butyl ether is more toxic than isopropyl ether by inhalation but less so by oral administration. Irritation of mucous membranes by the butyl derivative is greater than by isopropyl ether.

Industrial applications. A major application of ethers is as solvents for fats, oils, waxes, perfumes, resins, dyes, gums, and hydrocarbons. Since ethers are insoluble in water, they function well as extracting agents for organic compounds in such uses as separating acetic acid from dilute solutions. The solvent properties of mixtures of ethyl ether and alcohol with respect to cellulose nitrate are enhanced, with the result that this solvent mixture is used in the manufacture of collodion films and pyroxylin plastics, both of which are based on cellulose nitrate.

Use as solvents

In the chemical laboratory and to a certain extent in the chemical industry, ethers are used as anhydrous, inert media for the preparation of organic compounds by such special reactions as the Grignard and Wurtz–Fittig reactions. By changing the alkyl group of the ether, solvents with a wide range of boiling points and densities are obtained, thereby providing materials with specific physical properties for specific uses. Special precautions to prevent fires and explosions are always necessary.

Vapours of specific ethers, such as bis(2-chloroethyl)-ether, are toxic to insects, and these substances are therefore employed as insecticides, miticides, and soil fumigants. Ethyl ether is used to denature several alcohol formulations, making the product unsuitable for human consumption. Ethyl ether also is employed as a starter for diesel engines and as an entrainment agent (a substance that carries water with it on distillation) in the dehydration of ethyl and isopropyl alcohol.

Benzyl ether is a plasticizer for cellulose nitrate and acetate, improving their properties for use as plastics. Isopropyl ether is a good solvent for antiknock compounds, and it serves as a high-octane blending agent for gasoline.

Identification and analysis. The solubility of ethers in cold concentrated sulfuric acid is used to distinguish these compounds from the paraffinic hydrocarbons and the alkyl halides. The so-called Zeisel method, which cleaves ethers with hydriodic acid, measures the methoxy and ethoxy groups when the other hydrocarbon chain of the ether is higher than ethyl and is not tertiary.

Ethyl ether is the most important member of the ether family, and many of the methods used for its analysis are applicable to other ethers.

Ethyl ether is commercially available in various specified grades: anesthetic, absolute (for use as a chemical reagent), industrial, solvent, and synthetic. Analytical methods have been developed to measure small amounts of impurities, especially peroxides. (J.A.Mo.)

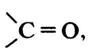
Aldehydes and ketones

Industrial uses

Aldehydes and ketones are families of similar organic chemical compounds, the molecules of which contain a special grouping of atoms called the carbonyl group. Largely because of the chemical properties of this group, aldehydes and ketones have many important industrial uses. Acetone, for example, is a solvent commonly employed in vast quantities in the production of paints, lacquers, and other substances. Acetaldehyde, like many other aldehydes and ketones, is a chemical intermediate, a substance converted by chemical reactions into other compounds widely employed in the production of plastics, dyes, and pharmaceuticals. In biological systems, many important compounds, such as the sugars and certain of the steroid hormones, contain aldehyde or ketone groups.

GENERAL CONSIDERATIONS AND PROPERTIES

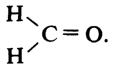
The carbonyl group, by which aldehydes and ketones are identified, consists of a carbon atom joined by two of its four available bonds to the two available bonds of an oxygen atom, and by the remaining two bonds to other carbon or hydrogen atoms. The molecular structure of the carbonyl group is written in chemical symbols as



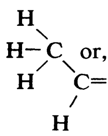
in which the letters C and O are the symbols for atoms of carbon and oxygen, respectively, and the lines represent chemical bonds between the atoms. As this partial formula indicates, the oxygen atom forms two chemical bonds (or in chemical terminology has a valence of two), and the carbon atom forms four bonds (valence, four), two of the valence bonds of carbon being utilized in forming the carbonyl bond with oxygen and two being available for joining to other atoms in the molecules of aldehydes and ketones.

Molecular structures

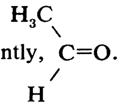
Aldehydes are distinguished from ketones in that they have at least one hydrogen atom (symbol H, valence one) joined directly to the carbon of the carbonyl group. The simplest aldehyde is formaldehyde, which is used in tanning, preserving, and embalming, and is a constituent of certain plastics. In its molecules two atoms of hydrogen are attached to the carbonyl carbon; its structural formula, therefore, is



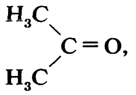
The next most complex aldehyde is acetaldehyde, in the molecule of which one carbon atom is joined to the carbonyl carbon, and the four valences of the second carbon are completed by joining three hydrogen atoms to it. The result is a molecule the structure of which can be designated by the formula



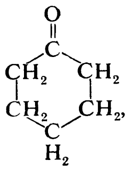
or, more conveniently,



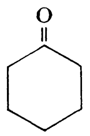
The CH_3 unit or group is called a methyl group. When two methyl groups are joined to the carbonyl carbon the result is acetone,



but acetone is a ketone, not an aldehyde, because there is no hydrogen atom attached directly to the carbonyl carbon. The compound called cyclohexanone is a cyclic ketone, its molecules consisting of a ring of six carbon atoms, one of which carries the doubly bonded oxygen of the carbonyl group. The structural formula of cyclohexanone is



or, in a notation customarily used by chemists,

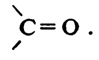


Aldehydes and ketones are so similar in their chemical and physical properties that they are generally classified together and often they are termed, collectively, carbonyl compounds, since it is the chemistry of the carbonyl group that identifies them. Each individual aldehyde or ketone has properties of its own, and all the members of both families readily undergo a wide variety of chemical reactions (a property which makes them highly suitable for use in chemical synthesis, such as the production of medicinal agents). Aldehydes generally are more chemically reactive than ketones. The simpler ones, such as formaldehyde and acetaldehyde, readily combine with themselves in repeated linkages to produce enormously long chains, containing tens of thousands of the simple molecule. Such macromolecules, known as polymers, can also be produced from the combination of aldehydes (especially formaldehyde) with other types of molecules. Many of these polymers are the familiar plastics of commerce; *e.g.,* Bakelite. Although some aldehydes and ketones occur in nature, including—in small amounts—acetone and acetaldehyde, most carbonyl compounds of commercial importance are prepared synthetically.

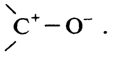
Function of the carbonyl group

The properties and behaviour of aldehydes and ketones are explained in large part by the nature of the carbonyl group. The addition of oxygen atoms to any element or compound is called oxidation; the removal of oxygen is called reduction. In a general definition of these reactions, oxidation means the removal of electrons from an atom or molecule, while reduction is the addition of electrons to an atom or molecule. Thus, an oxidizing agent removes electrons and, in the process of adding them to itself, becomes reduced. Similarly, a reducing agent becomes oxidized, since it donates electrons to the substance which it reduces. Actually, in many oxidation and reduction reactions, electrons are not physically removed or added but a compound is formed between the reactants. For example, if B oxidizes A some of A's electrons will also be held by B in the resulting molecule AB. An element or molecule may have more than one electron removed, either in a single oxidation process or in several steps. In the latter case, the substance will have more than one oxidation state, or number. The most completely oxidized state of the carbon atom is found in carbon dioxide, CO_2, which means that the carbon atom has all four of its valence electrons attached to the two oxygen atoms, bonding all three atoms together. The fully reduced state of carbon occurs in methane (CH_4), in which the four electrons of the hydrogen atoms are held by the carbon. The carbonyl group thus represents a partially oxidized state, and aldehydes and ketones are subject to either further oxidation or further reduction—a situation that accounts in part for their great chemical reactivity and for the wide variety of chemical reactions in which they take part.

Electronic structure of the carbonyl group. The chemical bonds that hold the nuclei of atoms together in molecular structures are the result of the motion of electrons around and between the constituent atoms. Thus, any group of atoms can be considered in terms of the electronic configuration that is associated with the bonded nuclei. Because it requires energy to disrupt any chemical bond, and because different kinds of bonds have different strengths, chemical reactions often can be predicted on the basis of known bond strengths. The electronic structure of the carbonyl group will clearly influence the physical and chemical characteristics of aldehydes and ketones. The carbonyl group may be pictured in terms of molecular orbitals that represent the "standing wave forms" of the electrons that are associated with particular bonds, in this case the carbon–oxygen double bond. The electronic structure of the carbonyl group can also be described in an alternate but chemically equivalent fashion as a resonance hybrid; *i.e.,* a combination of two structures, neither of which alone adequately depicts the molecule but which, taken together, represent its true structure. One of the resonance forms of the carbonyl group is the doubly bonded form already described:



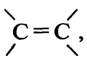
The other is one in which a positive charge appears on the carbon atom and a negative charge on the oxygen atom:



Hybrid state

The hybrid state of the actual carbonyl group is indicated in symbols by showing both forms connected by a double-headed arrow, as

The presence of plus and minus charges in one of the resonance structures suggests that the carbonyl group is highly polar (that is, that it shows an uneven distribution of electrons that makes the oxygen electrically negative and the carbon electrically positive). In contrast to a carbon–carbon double bond

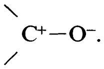


which is essentially nonpolar and weaker than should be expected from the combination of two carbon–carbon single bonds, the carbon–oxygen double bond is strongly polar, possessing more than twice the bond energy of a carbon–oxygen single bond.

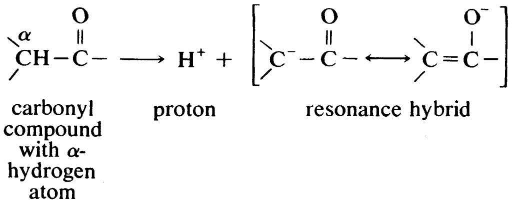
Physical properties. The polarity of the carbon–oxygen double bond is responsible for many of the physical properties of carbonyl compounds; for example, their dipole moment, which affects their behaviour in electric fields. The dipole moment of a substance is a measure of the uneven distribution of electrons that produces an asymmetry of electrical charges in the molecule. The highly polarized character of the carbonyl group introduces a certain degree of attraction between the molecules of carbonyl compounds; *i.e.,* since the oxygen of each carbonyl molecule carries more negative charge than the carbon atom, it will be attracted to the carbon end of other carbonyl molecules. This attraction in turn affects the melting points and boiling points of the compounds, both of which properties depend on the ease of separability of the molecules. Thus, the melting points of the simpler aldehydes and ketones are from 50° to 80° C (90° to 144° F) higher than those of hydrocarbons, compounds of comparable size, shape, and molecular weight but lacking these polar structural units. Although formaldehyde is a gas at ordinary temperatures, other low-molecular-weight carbonyl compounds are volatile liquids. The water solubility of simple carbonyl compounds is considerably enhanced because of the polar nature of the carbon–oxygen double bond (polar compounds generally being more water soluble than nonpolar compounds; see Table 22 for boiling points, melting points, and solubilities of common aldehydes and ketones).

Nucleophilic sources

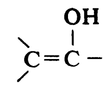
Chemical properties. The chemical properties, too, of aldehydes and ketones are largely determined by the particular electronic character of the carbonyl group. Thus, substances attracted to positive charges (called nucleophilic reagents because they contain electron-rich atoms) generally attack the carbonyl carbon, which bears a partial positive charge resulting from the contribution of the charged resonance form



Hydrogen atoms on the carbon atom immediately adjacent to the carbonyl, the so-called α-carbon, are easily removed from the molecule because the resulting negatively charged group (called an anion) is a resonance hybrid (see above) of two forms, a situation that provides extra stability to the structure.

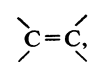


The effect of the resonance stabilization of the anion is to cause the α-hydrogen (hydrogen attached to the α-carbon) to be easily lost as a proton, that is, a transferable hydrogen nucleus represented by the symbol H^+, also called a hydrogen ion; in other words, the resonance stabilization makes the α-hydrogen acidic. A further consequence of the ease of loss of α-hydrogens is that carbonyl compounds bearing such hydrogens are easily converted to their so-called enol



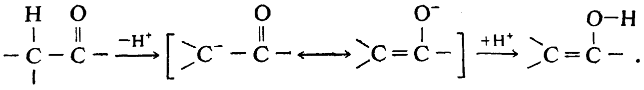
$$\overset{\text{OH}}{\overset{|}{>C=C-}}$$

forms (the name being derived from a combination of -ene, the ending for double-bond compounds,

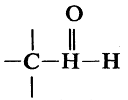


$$>C=C<,$$

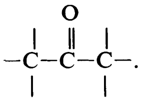
and -ol, the ending for hydroxyl compounds, —OH). The interconversion usually occurs through the resonance-stabilized anion (see above). The reaction is represented in equation form in the following manner:



Generally, aldehydes and ketones enter into a wide variety of chemical reactions involving the carbonyl group itself or the α-carbons (see below for specific examples). Aldehydes typically are more reactive than ketones, in part because the carbonyl carbon in aldehydes is more open to attacking groups because the hydrogen atom attached to it is less bulky than the carbon-containing groups attached to it in ketones, and this can be seen readily by comparing the formulas:



aldehyde



ketone

Nomenclature. The naming of organic compounds is complicated by the fact that any chain of carbon atoms may have branches or rings attached to it. When a dominant group, such as the carbonyl, is present, the basic design of the nomenclature is to chart complicated structures by reference to the dominant group. Many aldehydes and ketones are referred to by common names, and in such usage aldehydes are named by analogy with the carboxylic acids to which they are related; thus, formaldehyde from formic acid, acetaldehyde from acetic acid. Ketones, however, are commonly designated by the groups (called substituent groups) attached to the carbonyl carbon, along with the appellation ketone; *e.g.*, methyl ethyl ketone for the compound $CH_3COCH_2CH_3$, in which a methyl (CH_3) and an ethyl (C_2H_5) are the substituent groups, both being joined to the carbonyl carbon. In formal, or systematic, nomenclature of aldehydes and ketones, the longest chain carrying the carbonyl group is specified as the parent structure, and the name of the compound is derived from the name of the parent structure by use of the suffix -al if the compound is an aldehyde (propanal, CH_3CH_2CHO—common name, propionaldehyde from propionic acid) or -one if it is a ketone (propanone, CH_3COCH_3—common name, acetone). To indicate the location on the parent chain of any carbon atom, in order to specify where a substituent group may be attached, the carbon atoms are numbered, beginning with the aldehyde carbon. With the ketones, because the carbonyl atom may be anywhere in the chain, the numbering begins at that end of the chain that gives the ketone carbon the smallest possible number:

Common names and systematic names

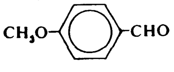
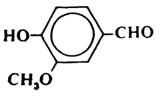
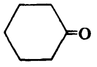
$$\overset{4}{C}H_3\overset{3}{C}H_2\overset{2}{C}O\overset{1}{C}H_3, \text{ not } \overset{1}{C}H_3\overset{2}{C}H_2\overset{3}{C}O\overset{4}{C}H_3.$$

In common terminology, the position of substituents is indicated by Greek letters, designating the carbon immediately adjacent to the carbonyl group with the Greek letter α, the second with β, and so on. Examples of these usages are: $ClCH_2CH_2CHO$, named 3-chloropropanal, or, commonly, β-chloropropionaldehyde; and

Designation of substituents

$$CH_3\underset{\underset{CH_3}{|}}{C}HCOCH_3,$$

Table 22: Aldehydes and Ketones

carbonyl compound	common name	systematic name	C=O infrared absorption frequency (cm⁻¹)	melting point (°C)	boiling point (°C)	water solubility	industrial uses
Aliphatic aldehydes							
HCHO	formaldehyde	methanal	1743*	−92	−21	very soluble	monomer for polymers (*e.g.*, Delrin), explosives (*e.g.*, Cyclonite)
CH_3CHO	acetaldehyde	ethanal	1752*	−121	21	soluble in all proportions	synthesis of acetic acid, acetic anhydride, ethyl acetate, acrylonitrile, crotonaldehyde, 1-butanol
Cl_3CCHO	chloral	trichloroethanal	1762†	−58	98	very soluble	synthesis of DDT
Aromatic aldehydes							
C_6H_5CHO	benzaldehyde	benzaldehyde	1704†	−26	178	slightly soluble	chemical intermediate, flavouring substance
	anisaldehyde	4-methoxybenzaldehyde	1700†	0	250	slightly soluble	perfume ingredient
	vanillin	4-hydroxy-3-methoxy-benzaldehyde	1700†	77–79	285	slightly soluble	flavouring substance
Aliphatic ketones							
CH_3COCH_3	acetone	2-propanone	1742*, 1715†	−95	56	soluble in all proportions	solvent, synthesis of mesityl oxide and methyl methacrylate
$CH_3COCH_2CH(CH_3)_2$	methyl isobutyl ketone	4-methyl-2-pentanone	1720†	−85	117	slightly soluble	solvent
	cyclohexanone	cyclohexanone	1710†	−45	155	slightly soluble	synthesis of monomer for polymers (*e.g.*, Nylon-6)
Aromatic ketones							
$C_6H_5COCH_3$	acetophenone	acetophenone	1686†	21	202	insoluble	chemical intermediate, perfume ingredient
$C_6H_5COC_6H_5$	benzophenone	benzophenone	1662†	48	306	insoluble	chemical intermediate
Unsaturated aldehydes and ketones							
$C_3CH = CHCHO$	crotonaldehyde	2-butenal	1685†	−74	104–5	very soluble	synthesis of 1-butanol
$(CH_3)_2C = CHCOCH_3$	mesityl oxide	4-methyl-3-penten-2-one	1690†	−53	130	soluble	synthesis of methyl isobutyl ketone

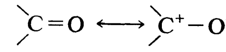
*In vapour phase. †In solution.

3-methyl-2-butanone or methyl isopropyl ketone. (A list of representative aldehydes and ketones of various types, with their common and systematic names, properties, and uses, is included in Table 22.)

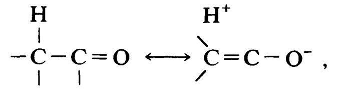
Principal methods of synthesis. Aldehydes and ketones can be synthesized by a wide variety of methods, some being generally useful for both classes of compounds and others being more specific in their application. These methods can be generally categorized as involving: (1) oxidation with cleavage of carbon–carbon bonds, (2) oxidation without such cleavage, (3) reduction (without cleavage), and (4) formation of new carbon–carbon bonds.

PRINCIPAL CHEMICAL REACTIONS

Further discussion of aldehydes and ketones requires a knowledge of electronic structure, chemical reactions, and vocabulary that cannot be provided here. The necessary background, however, can be found in detail in the articles ATOMS; CHEMICAL ELEMENTS; and CHEMICAL REACTIONS; and in the section *Organic chemical compounds.* The polarity of the carbonyl group, which reduces the electron density both at the carbonyl carbon,

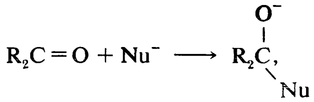


and at the α-hydrogen,



Addition and displacement reactions

promotes easy reaction with nucleophilic agents (groups that attack positively charged centres) and enhances the acidity of the α-proton. Two of the characteristic reactions of aldehydes and ketones, therefore, are addition processes in which a nucleophile adds to the carbonyl carbon—



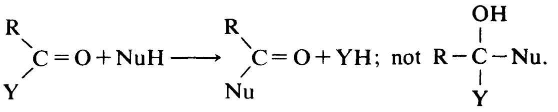
in which R is any hydrocarbon chain and Nu any nucleophilic agent—and displacement processes wherein the carbonyl compound loses an α-hydrogen and itself becomes a nucleophile—$RCH_2COR \rightleftharpoons R^-CHCOR + H^+$.

Major categories of reactions. *Addition of noncarbon nucleophiles to the carbonyl carbon.* The reactivity of nucleophiles based on atoms other than carbon (heteroatom nucleophiles) toward $R_2C{=}O$ increases according to the following sequence: halogen < oxygen < nitrogen < sulfur. Thus, the very weakly nucleophilic halide anions have virtually no tendency to become joined to aldehydes and ketones, and water molecules are added only to a few especially reactive ones. Acetone, for example, is hydrated only to a limited extent in aqueous solution, but hexafluoroacetone, strongly activated by the electron-withdrawing fluorine atoms, is almost completely converted to its hydrate under comparable conditions. Alcohols (oxygen compounds), amines (nitrogen compounds), and mercaptans (sulfur compounds), on the other hand, are added more effectively than water. The initial addition of such compounds is often followed by a dehydration step that may be either intramolecular (within the molecule) or intermolecular (between molecules). Intermolecular dehydration is observed following the acid-catalyzed addition of alcohols to ketones; the initially formed product, termed a hemiketal, reacts, via dehydration, with a second molecule of alcohol to give a second product called a ketal.

Addition of carbon nucleophiles to the carbonyl carbon. Carbon nucleophiles are especially reactive toward aldehydes and ketones, and this type of reaction forms the basis for many important synthetic methods. Among the carbon nucleophiles most frequently used for this purpose are hydrogen cyanide (HCN), Grignard reagents (RMgX, in which R is a hydrocarbon, Mg a magnesium atom, and X a halogen, usually chlorine or bromine), and other organometallic compounds, active methylene nucleophiles, including the nitroalkanes (RNO_2), and such highly polarizable substances as diazomethane (CH_2N_2) and certain ylides.

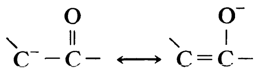
Comparison with carboxylic acid derivatives

A feature that generally differentiates aldehydes and ketones from carboxylic acid derivatives is the tendency of the latter to undergo nucleophilic displacement rather than addition; *i.e.,*



Alpha-trihalo ketones ($RCOCX_3$) and β-diketones ($RCOCH_2COR$), however, resemble carboxylic acid derivatives in this respect, the driving force for the displacement in these cases resulting from the special stability of the groups displaced, the $-CX_3$ and $-CH_2COR$ anions.

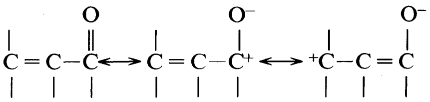
Displacement at the α-carbon. The capacity of the carbonyl group to attract and stabilize a negative charge facilitates the displacement of hydrogens attached to the α-carbon. The resulting resonance-stabilized (see above) anion



acts as a carbon nucleophile in a variety of reactions. Among others, it can react with: (1) molecular bromine, to give α-bromo compounds; (2) alkyl halides, to give α-alkylated compounds; and (3) other carbonyl compounds to give β-hydroxy compounds that may lose water to form unsaturated substances.

Conjugated systems

Conjugate addition. Electronic effects can be transmitted within organic molecules by way of conjugated systems, which are series of alternate single and double bonds between adjacent carbon atoms or between carbon atoms and atoms of other elements, such as oxygen. (Conjugated systems contain two or more directly-attached pi bonds—bonds in which the molecular orbitals are unsymmetrical around the bond axis.) Because aldehydes and ketones with a double bond between the α-carbon and the β-carbon (α–β-unsaturated carbonyl compounds) are conjugated systems, they exhibit this transmission of electronic effect. As a result, they exist as resonance hybrids (see above) in which one of the forms has a positive charge on the β-position rather than on the carbonyl carbon



$$\mathrm{C{=}C{-}\overset{O}{\overset{\|}{C}}} \longleftrightarrow \mathrm{C{=}C{-}\overset{O^-}{\overset{|}{C^+}}} \longleftrightarrow \mathrm{{}^+C{-}C{=}\overset{O^-}{\overset{|}{C}}}$$

As a result, nucleophiles may attack the β-carbon rather than the carbonyl carbon, leading to addition across the conjugated system, a process generally referred to as conjugate addition.

Oxidation–reduction reactions. Aldehydes and ketones are capable of gaining or losing electrons (*i.e.,* of undergoing, respectively, reduction or oxidation). Nucleophilic additions to aldehydes and ketones actually are examples of reduction, as is most explicitly shown when the hydride ion is the nucleophile (lithium aluminum hydride, sodium borohydride, and other metal hydrides are sources of the hydride ion) reducing aldehydes to primary alcohols ($RCHO \rightarrow RCH_2OH$) and ketones to secondary alcohols ($R_2C{=}O \rightarrow R_2CHOH$). When certain metals, such as magnesium, are used as reducing agents, the conversion of carbonyl compounds to dimeric (incorporating two molecules of the starting material) products may take place, a process called bimolecular reduction. Clemmensen reduction, one of the several methods available, employs amalgamated zinc (zinc–mercury alloys) in the presence of hydrochloric acid as the reducing medium. Another method consists of the conversion of the starting compound to a thioacetal or thioketal (sulfur-containing derivatives) followed by hydrogenolysis (cleavage with hydrogen) in the presence of a catalyst: $R_2C{=}O \rightarrow R_2C(SC_2H_5)_2 \rightarrow R_2CH_2 + 2C_2H_5SH$.

Clemmensen reduction

Aldehydes are considerably more susceptible than ketones to oxidation, since they are readily oxidized to carboxylic acids under conditions that have little effect on ketones. Under more strenuous conditions (such as heating with strong oxidizing agents), however, ketones can also be oxidized, by a process involving cleavage of carbon–carbon bonds associated with the carbonyl group. Thus, the action of hot nitric acid on cyclohexanone (see above and Table 22) cleaves the cyclohexane ring and gives the dicarboxylic acid, adipic acid, a substance used in the preparation of nylon. In the presence of strong base, an aldehyde lacking any α-hydrogen atoms (*e.g.,* HCHO, C_6H_5CHO, and R_3CCHO) undergoes a self-oxidation–reduction (the Cannizzaro reaction); treatment of benzaldehyde with strong aqueous sodium hydroxide yields a mixture of sodium benzoate, an oxidation product, and benzyl alcohol, a reduction product.

Polyfunctional aldehydes and ketones. The presence of additional functional groups within the molecule greatly influences the properties and behaviour of aldehydes and ketones. Such compounds, for example, frequently show altered infrared absorption characteristics (light absorption properties indicative of the electronic state of the carbonyl group), as shown in the Table. Conjugate addition (see above) is a direct consequence of the interaction of a carbonyl function with an α,β-unsaturated group within the molecule. Furthermore, carbonyl groups frequently are found to undergo direct chemical reaction with other functional groups within the same molecule. In many sugars, intramolecular addition of a hydroxyl group to the aldehyde or ketone function is known to occur. Glucose, for example, exists almost entirely in a cyclic, or ring, form.

Effect of one carbonyl on another

Similarly, in polycarbonyl compounds the effect of one carbonyl function upon another may frequently be observed. The enhanced reactivity of β-diketones toward substitution is a good example of such mutual influence. In some dicarbonyl compounds, an internal chemical reaction between the various portions of the molecule occurs. For example, intramolecular condensation is observed during the so-called Robinson procedure for preparing six-membered rings of carbon atoms, a method frequently employed in synthetic processes.

Synthesis with aldehydes and ketones. Because of (1) their ease of preparation by many different methods; (2) their relative stability and convenience; (3) their ability to react with protective groups, which insure their unreactivity during other chemical manipulations; and (4) their capacity for undergoing chemical reactions of almost unlimited variety, the aldehydes and ketones are nearly ideal chemical intermediates. There are, in fact, few multistep synthetic procedures that do not involve aldehydes or ketones at some stage in the process. To cite one example among many, a laboratory synthesis of longifolene—a hydrocarbon obtained from certain turpentines, involves carbonyl chemistry at nearly every stage of the process.

IDENTIFICATION AND USES

Character-ization of aldehydes

Analysis of aldehydes and ketones. Aldehydes and ketones as a group are readily characterized by their ability to form derivatives such as phenylhydrazones, oximes, and semicarbazones (see above). It was the use of phenylhydrazine to form characteristic crystalline derivatives of the sugars that permitted chemists to unravel many of the complexities of carbohydrate chemistry in the latter part of the 19th century. Characteristic, strong absorption bands in their infrared spectra (patterns of infrared light absorption) near 1720 cm^{-1} (reciprocal centimetres, a standard unit for light absorption measurements) also are useful in identifying aldehydes and ketones. Various substituents in the neighbourhood of the carbonyl group cause shifts in the absorption band that may be used to help identify the substituents. For example, as shown in the Table, aromatic and olefinic groups adjacent to the carbonyl function cause the absorption due to the carbonyl bond to move to the neighbourhood of 1680 cm^{-1}. Aldehydes are specifically characterized by their easy oxidation to carboxylic acids by such mild oxidants as silver ion (Tollens reagent) and cupric ion (Fehling reagent), neither of which has an appreciable effect on ketones. Nuclear-magnetic-resonance spectroscopy, a technique that measures the intensity of magnetic fields of atomic nuclei, also is useful in identifying aldehydes; in the spectra obtained with this technique, a characteristic band is found with aldehydes.

Physiologically active carbonyl compounds

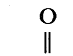
Aldehydes and ketones of commercial, physiological, and medicinal importance. A number of aldehydes and ketones are important industrial compounds and are manufactured on a large scale, in millions of tons. These and other aldehydes and ketones (see Table 22) are useful industrially as solvents, polymer components, flavouring agents, perfume ingredients, and chemical intermediates. Many compounds with important physiological properties contain carbonyl functions. Among those possessing an aldehyde group are retinene and pyridoxal phosphate. Retinene, obtained by oxidation of vitamin A_1, combines with a protein to form a pigment called rhodopsin, or visual purple, which plays a key role in the phenomenon of human vision. Pyridoxal phosphate is a coenzyme that, in association with a variety of enzymes, catalyzes certain essential life processes. The sugars are another important class of compounds in which either an aldehyde or ketone function is present. Among the common sugars are glucose (an aldohexose) and fructose (a ketohexose). Many compounds of medicinal value contain aldehyde or ketone functions. Many natural and synthetic steroid hormones bear aldehyde or ketone groupings. Molecules of the anti-inflammatory agent cortisone have three-ketone groups, and aldosterone, a salt-retaining hormone, has two ketone groups and an aldehyde group. (C.D.G.)

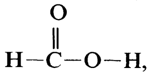
Carboxylic acids and their derivatives

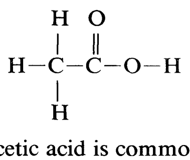
The carboxylic acids and their derivatives are organic chemical compounds, either prepared synthetically or found naturally in fats, vinegar, milk products, and fruit juices and flavours. They are used as solvents and in the preparation of other chemical compounds that are employed in the production of plastics, dyes, pharmaceuticals, and a large number of other products.

THE CARBOXYL GROUP

Structure

The molecules of the carboxylic acids contain a particular arrangement of carbon, oxygen, and hydrogen atoms, called a carboxyl group. It is this group that is responsible for the characteristic chemical behaviour of the carboxylic acids. In the carboxyl group, a carbon atom is joined to two oxygen atoms, one of which also is joined to a hydrogen atom. In chemical symbols, this structure of the carboxyl group is written  $-\overset{\overset{\displaystyle O}{\|}}{C}-O-H$, in which each line represents one chemical bond between the atoms. As indicated by this structure, the hydrogen atom forms only one bond (or, in chemical terminology, has a valence of 1); oxygen forms two bonds (valence, 2) and carbon four (valence, 4). In order to apportion the bonds properly, one of the oxygen atoms must be attached to the carbon by two bonds or, as it is commonly called, a double bond. The carbon atom has one unused bond left over after the carboxyl group is formed (as indicated by the line in front of the carbon symbol in the diagram above), and this bond must be used if the molecular structure is to be completed.

The simplest way that this can be done is to add a hydrogen atom, giving the simplest possible carboxylic acid, formic acid, the structure of which may be written  $H-\overset{\overset{\displaystyle O}{\|}}{C}-O-H$, or in shorter form as HCOOH or HCO_2H. Formic acid is used as a disinfectant and as an acidifying agent in textile and paper manufacturing.

Alternatively, a methyl group may be joined to the carbon of the carboxyl group to give acetic acid, with the structure  $H-\underset{\underset{\displaystyle H}{|}}{\overset{\overset{\displaystyle H}{|}}{C}}-\overset{\overset{\displaystyle O}{\|}}{C}-O-H$ or CH_3COOH. A dilute form of acetic acid is commonly known as vinegar.

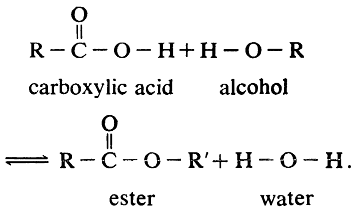
The structures of many other carboxylic acids may be made by joining other atoms to the methyl carbon in place of the hydrogen atoms. Such structures consist of chains of carbon atoms joined to one another in linear sequences, with side branches formed of other short chains and rings made when the chain doubles back and rejoins itself; each such structure represents the molecule of a different carboxylic acid. Often, for convenience, the symbol R is used to represent a general and unspecified arrangement of carbon and hydrogen atoms (a hydrocarbon unit) and the formula RCOOH is used as a general representation for all carboxylic acids.

The derivatives of the carboxylic acids all have molecular structures that contain variants of the carboxyl group. In general, these conform to the pattern $R-\overset{O}{\overset{\|}{C}}-Z$, in which R has the same meaning as above and Z is some atom other than carbon or hydrogen (*i.e.,* a heteroatom), usually oxygen, nitrogen, or a halogen (*i.e.,* one of the group of elements including fluorine, chlorine, bromine, and iodine). Often hydrogen (or other) atoms must be attached to the heteroatom to complete its valence bonds, and in these cases the Z represents a group of atoms. The carboxylic acids themselves conform to the general formula $R-\overset{O}{\overset{\|}{C}}-Z$, the Z in this case being the hydroxyl, or — OH, group.

Esters

Probably the most familiar class of carboxylic acid derivatives is the family of compounds known as esters. In esters, the heteroatom is oxygen, and a hydrocarbon unit is joined to it. In the methyl esters (the simplest of the family), the hydrocarbon unit joined to the oxygen atom is simply a methyl ($-CH_3$) group. Perhaps the most important ester of commerce is ethyl acetate, in which an ethyl ($-C_2H_5$) group is attached to the oxygen atom. Ethyl acetate is an ester of acetic acid, CH_3COOH. Its structural formula is $CH_3COOC_2H_5$. Ethyl acetate is used as a solvent, a flavouring and perfume ingredient, and a chemical intermediate (substance converted to other chemical compounds).

Esters can be formed from carboxylic acids by reaction with other organic compounds called alcohols under conditions that remove a molecule of water from the two constituents. This process is indicated in the following equation, which represents the chemical reaction that occurs:



$$\underset{\text{carboxylic acid}}{R-\overset{O}{\overset{\|}{C}}-O-H}+\underset{\text{alcohol}}{H-O-R} \rightleftharpoons \underset{\text{ester}}{R-\overset{O}{\overset{\|}{C}}-O-R'}+\underset{\text{water}}{H-O-H}.$$

In this equation, as in other chemical equations, the structural formulas of the starting materials are written to the left of the arrow and the products to the right. The double arrows indicate that the chemical reaction proceeding from left to right may be reversed, the products reacting to produce the reactants. In the structural formulas, written as above, the letters represent atoms of the various elements and the lines represent bonds between the atoms. Also, as above, R represents an unspecified hydrocarbon unit and R′ a second hydrocarbon unit, which may be the same as the first or may be different from it. As the general formula for esters, RCOOR′, suggests, this family of compounds is large, since both the R and R′ groups can be varied to include a great many kinds of hydrocarbon structures, as well as units with heteroatoms of various kinds substituted in them. Many simple esters, liquids with fruity flavours and odours, are used extensively in flavourings and perfumes. Some are naturally occurring.

Other derivatives

The other principal classes of carboxylic acid derivatives are the anhydrides, halides, peroxy acids, amides, hydrazides, and azides. The general formulas of these derivatives are given in Table 23. Each class can be considered to be formed from a carboxylic acid and a second compound by the same type of dehydration (removal of water) reaction that gives esters from carboxylic acids and alcohols. The second molecule that reacts with a carboxylic acid to give the derivative in each case is as follows: anhydride, a second molecule of carboxylic acid; halide, a halogen acid; peroxy acid, hydrogen peroxide; amide, an amine; hydrazide, hydrazine; and azide, hydrazoic acid. The preparation of the various derivatives, in actual practice, may not follow this idealized scheme since more convenient methods of synthesis have been worked out. Amides (particularly proteins) and esters are the most important carboxylic acid derivatives found in nature. Many of these derivatives, however, are synthesized for important industrial uses.

Typically, the carboxylic acids themselves are readily converted to the various classes of derivatives, and the derivatives in most cases can be converted into one another. Thus, the carboxylic acids and their derivatives comprise a family of closely related and readily interconvertible compounds. In many ways, the chemistry of each class is the chemistry of all.

THE CARBOXYLIC ACIDS AND DERIVATIVES AS A GROUP

Nomenclature. Because many of the more familiar carboxylic acids were first isolated from natural sources before the development of systematic chemical nomenclature, they are known by common names that reflect their origin. For example, it has long been known that ants secrete formic acid (Latin *formica,* "ant"), that acetic acid is present in vinegar (Latin *acetum,* "vinegar"), and that lactic acid is produced by the souring of milk (Latin *lac, lactis,* "milk"). Systematic names, which are generally preferred—except for the most common acids—because they are directly related to the chemical structure of the

Table 23: Formulas and Nomenclature of Carboxylic Acids and Their Derivatives

class	formula: $R-\overset{O}{\overset{\|}{C}}-Z$.	common name		systematic name	
	Z=	add to root*		add to root†	
		prefix	suffix	prefix	suffix
Carboxylic acid	$-OH$		-ic acid		-oic acid
Ester	$-OR'$	alkyl	-ate	alkyl	-oate
Anhydride	$-O\overset{O}{\overset{\|}{C}}OR$		-ic anhydride		-oic anhydride
Acyl halide	$-Cl$ $(-Br)$		-yl halide		-oyl halide
Peroxy acid	$-OOH$	peroxy- (or per-)	-ic acid	peroxy-	-oic acid
Amide‡ (substituted amide)	$-NH_2$ $(-NHR, -NR_2)$	(*N*-alkyl-, *N,N*-dialkyl-)	-amide	(*N*-alkyl-, *N,N*-dialkyl-)	-amide
Hydrazide	$-NHNH_2$		-ohydrazide		-ohydrazide
Azide	$-N_3$		-yl azide		-oyl azide

**E.g.,* acet-, propion-, etc. †*E.g.,* ethan-, propan-, etc. ‡Nitriles, another class of carboxylic acid derivative, may be considered to be dehydration products of amides. They have the general formula $R-C\equiv N$ and are named -onitrile.

acid, are derived from the name of the hydrocarbon with the same carbon skeleton. The name of the acid is derived by dropping the final "e" from the hydrocarbon name and replacing it with the ending "-oic acid." Thus, the name hexanoic acid is derived from that of the hydrocarbon hexane, to which it is formally related; this acid, however, usually is known by its common name, caproic acid (Latin *caper, capri,* "goat"). Systematic and common names of straight-chain saturated acids are given in Table 24.

Naming of derivatives

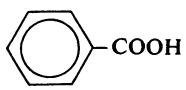
Methods of naming carboxylic acid derivatives vary. Suffixes or prefixes are used, or both, as indicated in Table 23, to indicate the type of derivative; the root of the acid's name gives the acid from which the compound is derived. As an example, in the naming of esters, the alcohol portion of the ester is given first, as a radical or partial molecule—such as ethyl from ethyl alcohol—followed by the root of the acid name—as acet- from acetic acid—followed by the suffix -ate. The ester formed from ethyl alcohol and acetic acid, then, is called ethyl acetate. When the systematic name of the acid (-oic) is employed, the ending used is -oate. The systematic name of ethyl acetate is ethyl ethanoate.

Table 24: Straight-chain Saturated Acids and Their Methyl Esters

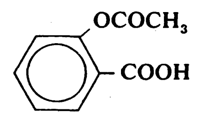
chain length	systematic name	common name	acid		methyl ester	
			melting point (°C)	boiling point* (°C)	melting point (°C)	boiling point* (°C)
1	methanoic	formic	8.4	101	−99.0	32
2	ethanoic	acetic	16.6	118	−98.1	57
3	propanoic	propionic	−20.8	141	−87.5	80
4	butanoic	butyric	− 4.3	164	−84.8	102
5	pentanoic	valeric	−33.8	186	−80.7	127
6	hexanoic	caproic	−2.0	205	−11.0	151
7	heptanoic	enanthic	−7.5	223	−55.7	172
8	octanoic	caprylic	16.5	239	−40.0	193
9	nonanoic	pelargonic	12.2	255	−34.3	214
10	decanoic	capric	31.5	270	−18	224
11	undecanoic	—	28.6	280	−11.3	250
12	dodecanoic	lauric	44.0	131^1	5.2	262
13	tridecanoic	—	44.5	140^1	6.5	90^1
14	tetradecanoic	myristic	58.0	149^1	19.1	114^1
15	pentadecanoic	—	53.5	158^1	18.5	154
16	hexadecanoic	palmitic	62.9	167^1	30.0	148^2
17	heptadecanoic	margaric	62.5	175^1	29.7	148^1
18	octadecanoic	stearic	71.5	184^1	39.1	156^1
19	nonadecanoic	—	69.4	—	38.5	191^4
20	eicosanoic	arachidic	77.0	204^1	46.4	188^2
21	heneicosanoic	—	75.2	—	—	207^4
22	docosanoic	behenic	80.0	—	54.0	206^2
23	tricosanoic	—	79.6	—	53.9	—
24	tetracosanoic	lignoceric	84.2	—	57.4	222^2
25	pentacosanoic	—	83.5	—	59.5	—
26	hexacosanoic	cerotic	88.5	—	63.5	237^2
27	heptacosanoic	—	87.6	—	64.6	—
28	octacosanoic	montanic	90.4	—	67.5	—
29	nonacosanoic	—	90.4	—	68.8	—
30	triacontanoic	melissic	93.6	—	71.5	—

*Boiling point at 760 millimetres unless at one, two, or four millimetres as indicated by superscript.

Classification. Carboxylic acids and their derivatives are classified in different ways, depending on the purpose of the classification. Many of the acids with straight carbon chains, particularly those with even numbers of carbon atoms, are constituents of fats and are commonly classed together as fatty acids. Generally the carboxylic acids are classed as aliphatic (a term that means an ordinary hydrocarbon chain) if the rest of the molecule (other than the carboxylic acid group) is aliphatic. The fatty acids, by this definition, are aliphatic; aliphatic acids frequently are classified further as saturated or unsaturated depending upon whether or not they contain unsaturation—that is, multiple bonds (double or triple) between the atoms of the carbon chain. When the carboxyl group is directly attached to an aromatic system—*e.g.,* a benzene ring—a carboxylic acid is, however, classed as an aromatic acid. Benzoic acid is the parent compound of the aromatic acids. Acetylsalicylic acid, or aspirin, the best known aromatic carboxylic acid, is derived from benzoic acid. The structural formulas of benzoic acid and acetylsalicylic acid are:



benzoic acid



acetylsalicylic acid (aspirin)

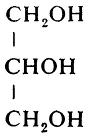
in which the hexagon containing a circle represents a benzene ring (that is, a circular structure of six carbon atoms each of which has one hydrogen atom attached to it).

Carboxylic acids are often categorized by the extra functional groups their molecules carry. Dicarboxylic and tricarboxylic acids are common. Carboxylic acids are classed as halo, keto, hydroxy, or amino acids when a halogen atom or a carbonyl, hydroxyl, or amino group is also part of the molecule.

Classification of derivatives

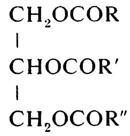
Generally carboxylic acid derivatives are placed in the same classes as the acids from which they are derived. Thus, aliphatic, aromatic, and various substituted esters, amides, and other derivatives are known. Exceptions occur when the acid-derived group makes up a relatively small part of the molecule. The derivative is then considered to be another type of substance altogether, modified by addition of the carboxylate grouping. Thus acetylsalicylic acid usually is considered a derivative of salicylic acid, not an ester of acetic acid.

Natural occurrence. Carboxylic acids occur widely in nature. In addition to acetic and formic acids, which are found in the free state, many of the fatty acids—straight-chain acids, saturated and unsaturated—occur as fats; that is, as esters of the trihydric alcohol glycerol (the term trihydric meaning that there are three hydroxyl or alcohol groups in each molecule of the substance). The structural formula of glycerol and a generalized formula of a fat derived from glycerol by the formation of ester bonds with three fatty acids are as follows:



$$\begin{array}{c} CH_2OH \\ | \\ CHOH \\ | \\ CH_2OH \end{array}$$

glycerol



$$\begin{array}{l} CH_2OCOR \\ | \\ CHOCOR' \\ | \\ CH_2OCOR'' \end{array}$$

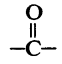
fat (R,R′, R″ = fatty acid side chains, saturated and unsaturated)

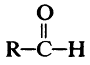
Fatty acids occur also as esters and amides of long-chain alcohols and amines. Free fatty acids are readily obtained from these sources by hydrolysis—cleavage of the ester or amide bond with addition of a single molecule of water. The free acids, however, are almost invariably obtained from such sources in the form of mixtures that are very complex and difficult to separate.

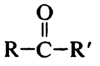
Substituted carboxylic acids also are found extensively in natural materials. Lactic acid occurs in sour-milk products and citric acid in fruits. These, other hydroxy acids, and many keto acids, are important metabolic products and exist in most living cells. Amino acids are constituents of proteins, from which they can be secured by hydrolysis. Halogen-substituted acids also occur, as do polycarboxylic acids and a number of aromatic acids.

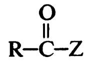
With the exception of esters and amides, few carboxylic acid derivatives appear in natural products. Esters, with their fruity and fragrant odours, are common constituents of plant flavours and essences. The carboxylic acids and their derivatives are all readily available by synthetic procedures (see below), often of such simplicity that even naturally occurring substances are prepared synthetically rather than isolated from natural sources.

General chemical and physical properties. The carboxyl group, in a sense, is a special form of the carbonyl group, the doubly bonded carbon–oxygen function characteristic of aldehydes and ketones. Structural representations of the carbonyl group and the generalized formulas of aldehydes, ketones, and carboxylic acids (and derivatives) are all given here for purposes of comparison:

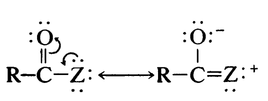

$$-\overset{\overset{O}{\|}}{C}-$$
carbonyl group


$$R-\overset{\overset{O}{\|}}{C}-H$$
aldehyde


$$R-\overset{\overset{O}{\|}}{C}-R'$$
ketone


$$R-\overset{\overset{O}{\|}}{C}-Z$$
carboxylic acid and derivatives

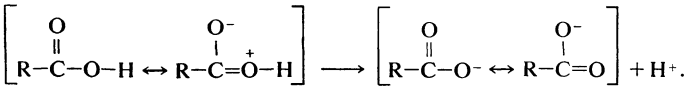
In the molecules of aldehydes and ketones, the carbonyl group is attached directly to carbon or hydrogen atoms, neither of which exerts a major influence on the chemical behaviour of that carbonyl group. In the carboxylic acids and their derivatives, however, the carbonyl carbon is joined directly to an atom with one or more pairs of electrons over and above those employed in the bond with the carbon atom. These so-called unshared pairs of electrons are able to interact with the electrons of the carbonyl group, thereby greatly modifying the chemical properties of that group and of the heteroatom as well. The nature of this interaction usually is represented in terms of the following diagrams:



In these diagrams the dots represent single electrons, situated in pairs; the curved arrows indicate a tendency for the electron pairs, or the bonds (which are equivalent to pairs of electrons), to migrate as shown; and the plus and minus signs stand for local electric charges brought about by the movement of the electrons. The double-headed arrow between the two representations implies that the complete structure of the molecule is a resonance hybrid of these two structures—that is to say, that the molecular structure incorporates aspects of both forms but duplicates neither. When such resonance forms exist for a molecule it is said to be stabilized by resonance. In this case, the most important feature of the electronic interaction represented by the resonance forms is that it gives this modified carbonyl group new chemical properties of its own, which justify classifying the group as a new and unique structure; *i.e.,* a carboxyl group.

As the name indicates, the chief characteristic of carboxylic acids is their acidity. As a class the carboxylic acids are more acidic than most other organic compounds containing hydroxyl (—OH) groups, but they are generally much weaker acids than the familiar mineral acids, such as nitric, sulfuric, and hydrochloric acids. The acidity of the carboxyl group results from the loss of a proton or hydrogen ion (that is, a hydrogen atom bearing a positive charge, as indicated by the symbol H^+), leaving behind a carboxylate anion—a negatively charged unit derived from the carboxyl group.

The carboxylate anion, like the carboxyl group, exists as resonance hybrid (see above) of several forms, two of which are shown below for each structure:


$$\left[R-\overset{\overset{O}{\|}}{C}-O-H \leftrightarrow R-\overset{\overset{O^-}{|}}{C}=\overset{+}{O}-H\right] \longrightarrow \left[R-\overset{\overset{O}{\|}}{C}-O^- \leftrightarrow R-\overset{\overset{O^-}{|}}{C}=O\right]+H^+.$$

carboxylic acid carboxylate anion

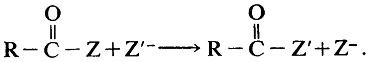
The basis for acidity of carboxylic acids

It is significant that the two resonance forms shown for the carboxylate anion are symmetrical and both carry a negative charge (as shown by the minus signs on the symbols for the oxygen atoms); this situation, which produces a structure of low energy content, means that the carboxylate anion is a relatively stable structure. In the carboxylic acid molecule itself, however, the resonance forms (as shown) are not symmetrical and one of them carries a separation of charge in its structure. This structure is less favourable (shows less resonance stabilization) than the symmetrical, equally charged structures of the carboxylate anion. As a result of this disparity in resonance stabilization, the carboxylic acid readily gives up a proton—acts as an acid—to produce the anion, which is favoured by its extra resonance energy.

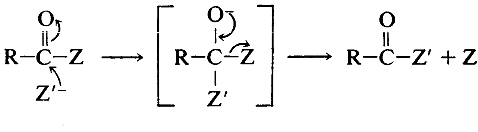
The nature of the R group of the carboxylic acid has an effect on the acidity of the carboxyl group by making it more or less easy for the proton to depart. Generally speaking, electron-attracting groups increase acidity by withdrawing electrons from the vicinity of the proton, thereby facilitating its departure; the opposite is true of electron-repelling groups. In many of the carboxylic acid derivatives, the carboxyl group has been modified, with the result that the proton in question is no longer present, and such substances are not acidic.

Chemical properties

The chemical reactions undergone by the carboxylic acids and their derivatives can be put into two broad categories. The first is a nucleophilic (positive-charge-seeking) reaction directed at the carbonyl carbon atom. The overall effect of this type of reaction is the displacement of one group, Z, which is capable of carrying a negative charge by another, Z′. The reaction is customarily represented by the following equation:

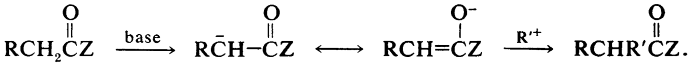

$$R-\overset{\overset{O}{\|}}{C}-Z+Z'^- \longrightarrow R-\overset{\overset{O}{\|}}{C}-Z'+Z^-.$$

The reaction proceeds, however, by way of attack on the carbonyl carbon as shown in the representation below:



in which the movement of electron pairs (or bonds) is indicated by curved arrows and the structure enclosed within brackets is an intermediate substance that is not isolated. It is by reactions of this type that carboxylic acid derivatives are interconverted, one variety being readily prepared from another under the proper chemical conditions.

Secondly, carboxyl compounds undergo the sort of reaction undergone by typical carbonyl compounds (aldehydes and ketones), in which the substituents are introduced into the so-called alpha position (the carbon atom adjacent to the carbonyl or, in this case, carboxyl group). Such reactions proceed by way of an intermediate resonance-stabilized anion (an anion being a negatively charged unit), which subsequently reacts with a positively charged unit. The latter is incorporated into the final structure. The reaction occurs as shown in the following equation:

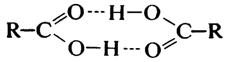

$$RCH_2\overset{\overset{O}{\|}}{C}Z \xrightarrow{\text{base}} R\overset{-}{C}H-\overset{\overset{O}{\|}}{C}Z \longleftrightarrow RCH=\overset{\overset{O^-}{|}}{C}Z \xrightarrow{R'^+} RCHR'\overset{\overset{O}{\|}}{C}Z.$$

resonance-stabilized anion

Esters resemble carbonyl compounds more than the carboxylic acids—or the other derivatives—do, and, as a result, esters are more prone to undergo substitution reactions in this category.

Physical properties

The members of the carboxylic acid family with low molecular weight are liquids with boiling points appreciably higher than those of hydrocarbons of comparable molecular weight (see Table 24). This is because of extensive hydrogen-bond formation, a type of secondary association between molecules, which makes the acids behave as dimers (two molecules held together as one):



As would be expected from compounds whose molecules contain hydroxyl groups, the lower members of the series are extremely water-soluble—the hydroxyl group (—OH) being structurally similar to the water molecule (HOH) and therefore prone to solubility in it. This property diminishes as the non-carboxylic acid portion of the molecules increases in size, and acids containing nine or more carbon atoms are insoluble. Because of their solubility and acidity, members of the series containing few carbon atoms—such as acetic acid, found in vinegar—have a sour taste. The carboxylic acids with one to three carbons have sharp odours, whereas those with four to eight carbons have rank, disagreeable odours. Rancid butter and strong cheese contain the latter carboxylic acids.

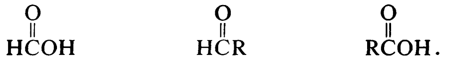
Many carboxylic acid derivatives are less associated (that is, their molecules are held together less strongly) than the free acids. This is especially true of the esters and halides, whose boiling points are much lower than those of the acids from which they are derived (see Table 24). Typical esters of lower molecular weight are volatile liquids.

THE MAJOR CLASSES OF CARBOXYLIC ACIDS

Saturated acids. Formic acid, the simplest member of the carboxylic acid family, was first prepared by distilling ants with water. The stinging nettle owes its irritating quality to the formic acid content of the juice. On a commercial scale, the acid is prepared by adding carbon monoxide under pressure to hot sodium hydroxide. The free acid is liberated by careful reaction with sulfuric acid. The series of reaction is shown below:

$$CO \xrightarrow{NaOH} HCOONa \xrightarrow{H_2SO_4} HCOOH$$

in which the symbols NaOH and H_2SO_4 written above the arrows indicate that the successive reactions occur in the presence of sodium hydroxide and sulfuric acid, respectively. In addition to its acidic properties, formic acid shows pronounced reducing activity (reduction being the chemical opposite of oxidation), a quality not shared by other carboxylic acids. The reducing action stems from the fact that formic acid is in a sense an aldehyde as well as an acid, as shown in the following structural formulas:



$$H\overset{\overset{O}{\|}}{C}OH \qquad H\overset{\overset{O}{\|}}{C}R \qquad R\overset{\overset{O}{\|}}{C}OH.$$

formic acid aldehyde carboxylic acid

Formic and acetic acids

Acetic acid, the next higher homologue—that is, the compound with one more carbon atom in its molecule—can be produced by oxidation of ethyl alcohol and is present as a dilute solution in vinegar, where it is formed by bacterial oxidation of dilute alcoholic liquors. Acetic acid may also be isolated from the liquid called pyroligneous acid, which is formed by the destructive distillation of wood. The most important commercial preparation of acetic acid involves the catalytic atmospheric oxidation of acetaldehyde, itself prepared from acetylene. These reactions are shown below:

$$CH{\equiv}CH \xrightarrow[HgSO_4]{\text{dilute } H_2SO_4} CH_3CHO \xrightarrow[Mn(CH_3COO)_2]{\text{air}} CH_3COOH.$$

Acetic acid is extensively employed in the preparation of metal salts and in the form of acetyl chloride or acetic anhydride in the production of esters and amides. Aspirin, the acetic ester of salicylic acid, is a good example of a commercially important ester prepared from acetic acid. Acetic acid is also recognized as an important substance in the biosynthesis of many biologically important materials. In this case it chiefly appears as an acetyl group bound to coenzyme A—*i.e.,* a molecule necessary for the action of certain enzymes. Acetyl-S-coenzyme A takes part in the biological production of fatty acids, many aromatic compounds, and isoprenoids and steroids. The natural fatty acids have an even number of carbon atoms in their molecules because they are produced from acetyl-S-coenzyme A, which adds carbon atoms to the molecules two at a time.

With increasing numbers of carbon atoms per molecule, the carboxylic acids have increasing possibilities for alternate arrangements of the atoms within the molecules (isomerism), and many branched-chain acids, both natural and synthetic, are known. The straight-chain compounds are, nevertheless, more common, especially those with even numbers of carbon atoms in the chain, and these make up the great percentage of the fatty acid content of common fats.

Carboxylic acids and their derivatives

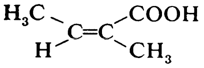
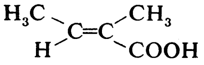
The lower fatty acids (with four to 10 carbon atoms) occur mainly in milk fats and a few seed fats. Cow's milk contains 10 percent (on a molecular-weight basis) of butyric acid (a four-carbon acid) with smaller amounts of the acids containing six, eight, 10, and 12 carbon atoms; sheep's-milk and goat's-milk fats contain these same acids, with capric (containing 10 carbon atoms) in greatest amount (up to 10 percent). Capric acid also is a major component (60 percent) in the seed fat of the elm, and caprylic and capric acids (each 5–10 percent) accompany the higher proportion of lauric acid (12 carbon atoms) in coconut oil and, to a lesser extent, in palm-kernel oil. Lauric acid and myristic acid (14 carbon atoms) occur extensively in seed fats of the Lauraceae (laurel family) and Myristicaceae (nutmeg family) respectively. Lauric acid, for example, can be obtained from cinnamon oil (80–90 percent), coconut oil (45–50 percent), or palm-kernel oil (45–55 percent) and myristic acid from nutmeg butter (60–75 percent) or from the more readily available coconut and palm-kernel oils (15–18 percent). Palmitic acid (16 carbon atoms), the most widely occurring of the saturated acids, is present in almost every fat examined. Useful sources include cottonseed oil (22–28 percent), palm oil (35–40 percent), and Chinese vegetable tallow (60–70 percent).

Palmitic and stearic acids

Stearic acid (18 carbon atoms) is less common than palmitic acid. It occurs in most vegetable fats but is a major component in only a few, such as cocoa butter (35 percent). It is present in most animal fats but attains major proportions only in the tallows of ruminants (30 percent). Stearic acid is conveniently obtained by hydrogenation of readily accessible unsaturated fats.

Still higher fatty acids are less common but occur in a few seed fats or in some waxes as esters of long-chain alcohols. Acids with odd numbers of carbon atoms in their chains are even rarer, but modern methods of investigation have revealed their presence, often as trace components, in many fats. Acids with hydrocarbon chains containing short branches (side chains of carbon atoms) or rings (circular sequences of carbon atoms) also are uncommon. Those that are found include a series of acids with one or more methyl groups in various positions along the chain, as well as acids with cyclopropane groups. Many branched-chain acids have been synthesized.

Unsaturated acids. Although most of the important unsaturated acids (acids with multiple bonds) are long-chain compounds obtained from fats, some shorter chain acids, such as acrylic, crotonic, methacrylic, and the isomeric angelic and tiglic acids, are also important. The formulas of these acids are shown below:

$CH_2{=}CHCOOH$	acrylic acid
$CH_3CH{=}CHCOOH$	crotonic acid
$CH_2{=}C(CH_3)COOH$	methacrylic acid
	angelic acid
	tiglic acid

Acrylic acid polymers

Acrylic acid, employed as its methyl ester in the production of polymers known as acrylates, is prepared on a large scale from acetylene, carbon monoxide, and water in the presence of a nickel catalyst, as well as by hydrolysis of the substance acrylonitrile ($CH_2{=}CHCN$), an even more important monomer.

Methacrylic acid, in the form of its methyl ester, is also used as a monomer, furnishing Perspex or Plexiglas when polymerized. Crotonic acid exists in *cis* and *trans* forms, which differ in the arrangement of substituents about the double bond. Only the *trans* isomer occurs naturally, in croton oil. Angelic and tiglic acids also are a pair of *cis* and *trans* isomers, both of which occur naturally.

Natural unsaturated acids number more than 200 and are subdivided into three groups: (1) acids with only one unsaturated centre; (2) conjugated (conjugation being an alternation of single and double bonds), polyunsaturated acids; (3) multiply unsaturated, or polyunsaturated, acids that are nonconjugated.

Monounsaturated acids

Apart from two acetylenic (triple-bonded) acids, tariric and stearolic, the natural monounsaturated acids are olefinic (double bonded) and almost entirely *cis* isomers.

Oleic acid, with 18 carbon atoms and a double bond between the ninth and tenth carbon atoms in the chain, is the most common of all fatty acids and the prototype for all monoolefinic acids:

$$CH_3(CH_2)_7CH = CH(CH_2)_7COOH \qquad \text{oleic acid.}$$

More than 90 monoenoic acids, containing 10 to 30 carbon atoms, have been identified in lipid sources. Nature appears to have a preference for certain chain lengths, especially those of 16 and 18 carbons, and for particular positions for the unsaturation, especially between C_9 and C_{10}. These circumstances are a natural consequence of the enzymatic oxygen-dependent desaturation process whereby double bonds are most commonly introduced into fatty acids between carbons 9 and 10, and of the readiness by which natural acids undergo chain elongation by one or more two-carbon units at the carboxyl end of the molecule. In addition to these natural acids, many isomers that do not occur in nature have been synthesized.

The monounsaturated acids show the properties expected of olefins, in addition to those of carboxylic acids. As acids, they form salts, are reduced to primary alcohols, and are converted directly or indirectly to the various types of acyl compounds listed in Table 23. As unsaturated compounds they undergo hydrogenation (that is, the addition of hydrogen to the multiple bond) with hydrogen and a catalyst or with hydrazine; hydroxylation (introduction of hydroxyl groups) with potassium permanganate, osmium tetroxide, performic acid, or peracetic acid; and halogenation (introduction of halogen atoms) with chlorine or bromine or with iodine monochloride. Vigorous oxidation by potassium permanganate or ozone leads to breaks in the carbon chains (chain fission). Atmospheric oxidation first gives unsaturated hydroperoxides (molecules containing two joined oxygen atoms), followed by a variety of secondary reaction products. Among these are volatile short-chain compounds (aldehydes and other substances) responsible for both the acceptable and unacceptable flavours that develop in fat-containing foods.

Polyunsaturated acids

More than 90 nonconjugated, polyunsaturated acids occur naturally. Most are acids with 16, 18, 20, or 22 carbons and two to four double bonds, although more highly unsaturated members are also known. Unsaturation is almost entirely *cis* in the compounds and usually is methylene interrupted; *i.e.,* with successive double bonds separated by single methylene ($-CH_2-$) groups. These structural features are clearly apparent in linoleic acid, linolenic acid, and arachidonic acid, the most important members of this class, the structures of which are shown below:

linoleic acid

$$CH_3(CH_2)_4CH = CHCH_2CH = CH(CH_2)_7CO_2H$$

linolenic acid

$$CH_3CH_2CH = CHCH_2CH = CHCH_2CH = CH(CH_2)_7CO_2H$$

arachidonic acid

$$CH_3(CH_2)_4CH = CHCH_2CH = CHCH_2CH = CHCH_2CH = CH(CH_2)_3CO_2H.$$

Interest in acids of this type has increased since the recognition that some of them, called essential fatty acids, are required for the maintenance of normal animal growth, reproduction, and permeability of the skin. The particular requirement is for methylene-interrupted polyolefinic acids, with unsaturation beginning at the sixth carbon from the methyl end of the chain.

Arachidonic acid is the most effective of the fatty acids for meeting the essential fatty acid requirements. Many authorities consider it likely that essential fatty acids are converted to other compounds before fulfilling their as yet unknown biological function.

Of possible significance in this regard is the discovery of a series of compounds called prostaglandins, which cause certain muscles to contract and lower the blood pressure. These compounds are produced in vivo from essential fatty acids such as arachidonic.

Conjugated polyunsaturated acids are much less common. About 40 are known to occur naturally, almost all of them being 18-carbon acids of plant origin. The best known is eleostearic acid from tung oil.

Substituted acids. Carboxylic acids, particularly those with long hydrocarbon chains, may contain one or more other groups, such as halo, hydroxy, alkoxy, epoxy, or keto groups. The several functional groups usually operate independently, and most substituted acids show only those properties that would be expected from the two functional groups. Sometimes, however, the two functional groups interact to give new properties.

Halogen-substituted acids

Acids substituted with chlorine in the alpha position are prepared by the reaction of chlorine or sulfuryl chloride on an acyl chloride or on the acid itself under reaction conditions that produce an acyl chloride. For example:

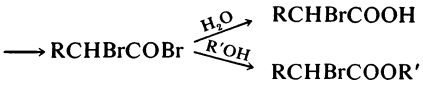
$$RCH_2COCl \xrightarrow{Cl_2,\ I_2} RCHClCOCl \xrightarrow{H_2O} RCHClCOOH$$

$$RCH_2COOH \xrightarrow{Cl_2,P} [RCH_2COCl] \longrightarrow RCHClCOCl \xrightarrow{H_2O} RCHClCOOH$$

$$RCH_2COOH \xrightarrow[SOCl_2]{SO_2Cl_2} [RCH_2COCl] \longrightarrow RCHClCOCl \xrightarrow{H_2O} RCHClCOOH.$$

The corresponding bromo acids are prepared by reactions with bromine and phosphorus, phosphorus trihalide, or thionyl chloride:

$$RCH_2COOH \xrightarrow{Br_2,\ P} [RCH_2COBr]$$



$$\longrightarrow RCHBrCOBr \begin{cases} \xrightarrow{H_2O} RCHBrCOOH \\ \xrightarrow{R'OH} RCHBrCOOR' \end{cases}$$

Iodo acids are best prepared by halogen exchange between potassium iodide and the chloro acid or bromo acid.

Acids substituted with halogen in other positions are produced by standard procedures for preparing alkyl halides, such as the addition of halogen acid to a double bond, or by the conversion of a hydroxyl group to halogen. Examples are:

$$RCH{=}CHCOOH \xrightarrow{HBr} RCHBrCH_2CO_2H$$

$$RCH(OH)(CH_2)_4COOH \xrightarrow{SOCl_2} [RCHCl(CH_2)_4COCl] \xrightarrow{H_2O} RCHCl(CH_2)_4COOH$$

The reactions of halogen-substituted acids and other acyl derivatives include dehydrohalogenation, nucleophilic substitution, and the formation of certain organometallic derivatives:

$$CH_2BrCH_2(CH_2)_nCOOH \xrightarrow{\text{pyridine}} CH_2{=}CH(CH_2)_nCOOH$$

$$RCHBrCOOH \xrightarrow[\text{of } Z^-]{\text{HZ or source}} RCHZCOOH$$

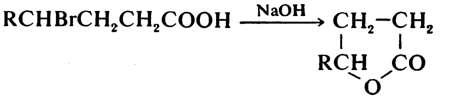
$$(Z{=}-NH_2, -OH, -CN, \text{etc.})$$

$$RCHBrCOOCH_3 \xrightarrow{Zn} RCH(ZnBr)COOCH_3 \xrightarrow{R'CHO} R'CH(OH)CH(R)COOCH_3 \xrightarrow[\text{hydrolysis}]{-H_2O} R'CH{=}C(R)COOH.$$

The presence of the carboxyl group influences the reactions of the 2-, 3-, and 4-halo acids so that each of these furnishes a different save of product when it reacts with alkalies:

$$RCHBrCOOH \xrightarrow{NaOH} RCH(OH)COOH \qquad \textbf{(2-hydroxy acid)}$$

$$RCHBrCH_2COOH \xrightarrow{NaOH} RCH{=}CHCOOH \qquad \textbf{(}\boldsymbol{\alpha, \beta}\textbf{-unsaturated acid)}$$



(γ-lactone)

Among the few naturally occurring halo acids is fluoroacetic acid, a highly toxic compound that occurs as the potassium salt in the African shrub *Dichapetalum cymosum.* Compounds that undergo degradation in vivo to fluoroacetic acid are also highly toxic, and 18-fluorooleic acid, for example, is the poisonous principle in certain unusual seeds.

Hydroxy and amino acids

Hydroxy acids can be prepared from the appropriate carboxylic acid intermediates by standard methods of preparing hydroxy compounds, or from the appropriate hydroxy compounds by methods for producing carboxylic acids. Some of the more important general methods include (1) the hydrolysis of halo acids, (2) the hydrolysis of cyanohydrins, (3) the reaction with α-bromo esters in the presence of zinc, and (4) the reduction of keto esters or half esters of dibasic acids:

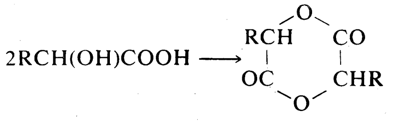
$$(1)\quad RCH_2COOH \xrightarrow{Br_2,\,P} RCHBrCOOH \xrightarrow{NaOH(aq)} RCH(OH)COOH$$

$$(2)\quad \left\{\begin{aligned} &RCHO \xrightarrow{HCN} RCH(OH)CN \xrightarrow{\text{hydrolysis}} RCH(OH)COOH \\ &RCH{=}CH_2 \xrightarrow{HOCl} RCH(OH)CH_2Cl \\ &\xrightarrow{KCN} RCH(OH)CH_2CN \xrightarrow{\text{hydrolysis}} RCH(OH)CH_2COOH \end{aligned}\right.$$

$$(3)\quad RCHO \xrightarrow{BrCH_2COOC_2H_5} RCH(OH)CH_2COOH$$

$$(4)\quad \left\{\begin{aligned} &RCO(CH_2)_nCOOCH_3 \xrightarrow[\text{hydrolysis}]{NaBH_4;} RCH(OH)[CH_2]_nCOOH \\ &HOOC(CH_2)_nCOOCH_3 \xrightarrow{SOCl_2} ClCO(CH_2)_nCOOCH_3 \\ &\xrightarrow[\text{hydrolysis}]{NaBH_4;} HOCH_2(CH_2)_nCOOH \end{aligned}\right.$$

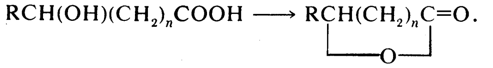
The 2-, 3-, 4-, and 5-hydroxy carboxylic acids differ in their behaviour on heating and give, respectively, a lactide, α, β-unsaturated acid, a γ-lactone, and a δ-lactone:



lactide

$$RCH(OH)CH_2COOH \longrightarrow RCH{=}CHCOOH$$

α, β-unsaturated acid



$$RCH(OH)(CH_2)_nCOOH \longrightarrow RCH(CH_2)_nC{=}O.$$

lactone

There are many naturally occurring hydroxy acids, including glycolic, lactic, and ricinoleic:

$HOCH_2COOH$	$CH_3CH(OH)COOH$
glycolic acid	lactic acid

$$CH_3(CH_2)_5CH(OH)CH_2CH{=}CH(CH_2)_7COOH$$

ricinoleic acid

Ricinoleic is the major acid present in castor oil and provides a source of other useful chemicals prepared on an industrial scale. These include the diene glycerides resulting on dehydration of castor oil, as well as the products of pyrolysis—heptanal and 10-undecenoic acid—and alkali fusion—2-octanone, 2-octanol, 10-hydroxy-decanoic acid, and sebacic acid.

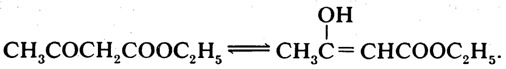
The α-amino acids, $RCH(NH_2)COOH$, are of great importance since they are the monomeric units that make up the peptides and proteins. Twenty different amino acids commonly occur in these materials.

Keto acids

Carboxylic acids that also carry carbonyl groups (aldehydes or ketones) usually can be prepared by oxidation of the corresponding hydroxy acids. The β-keto compounds (in the form of their esters) are of considerable importance and are prepared by special condensation procedures.

Among the more important shorter chain acids bearing extra carbonyl groups are glyoxylic acid (OHCCOOH), pyruvic acid ($CH_3COCOOH$), levulinic acid ($CH_3COCH_2CH_2COOH$), and especially acetoacetic acid (CH_3COCH_2COOH).

Ethyl acetoacetate and most β-keto esters exhibit tautomerism—that is, the esters exist as equilibrium mixtures of two compounds and exhibit the properties of both. The keto form, which preponderates in most cases, contains a normal carbonyl group; the other form, called the enol form, bears an alcoholic group attached directly to one of the carbons of a double bond. The enol form is usually present in small amounts, but it can be detected by its reactions. As the enol form is consumed in a reaction, more is generated from the keto form by re-establishment of the equilibrium:

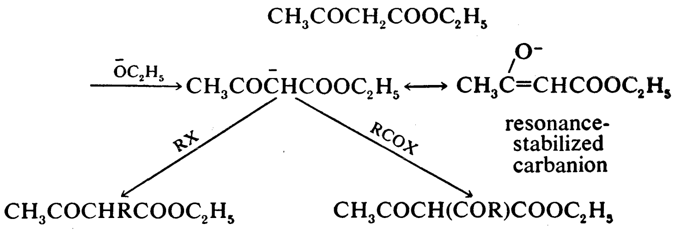


$$CH_3COCH_2COOC_2H_5 \rightleftharpoons CH_3\overset{OH}{C}{=}CHCOOC_2H_5.$$

keto form enol form

Typical reactions of the keto form include reactions with hydrogen cyanide and with hydroxylamine to give cyanohydrins and oximes respectively. On the other hand, reaction with diazomethane to give an *O*-methyl ether and with ferric chloride to yield a coloured complex are considered to be typical of the enolic form.

Of major importance chemically (because it is the basis for many chemical reactions) is the fact that the methylene ($-CH_2-$) group of a β-keto ester, which is flanked by two electron-withdrawing groups ($-CO-$ and $-COOC_2H_5$), can give up one of its protons to a strong base to yield a resonance-stabilized carbanion. Intermediates of this type react readily with alkyl and acyl halides to give alkylated or acylated derivatives of the original β-keto ester, as follows:



The monoalkylated and monoacylated compounds formulated in the above sequence also contain a second hydrogen atom, which can be replaced by a second alkyl or acyl group, not necessarily the same as the first. These alkylated or acylated derivatives yield valuable products after hydrolysis, a reaction that can follow two different courses:

1. Under mild conditions—*i.e.,* with dilute acid or alkali—the ester function is hydrolyzed in the usual way, but the β-keto acid so formed is unstable and decarboxylates readily to give a ketone:

$$CH_3COCH_2COOC_2H_5 \longrightarrow CH_3COCHRCOOC_2H_5$$

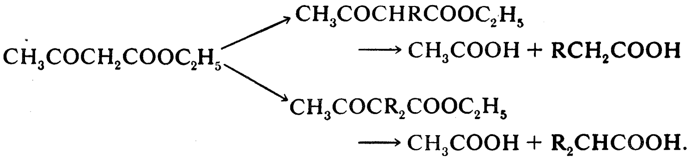
ethyl acetoacetate mono-alkylated derivative

$$\longrightarrow CH_3COCHRCOOH \longrightarrow CH_3COCH_2R.$$

free acid ketone

Reactions of this type are used to prepare monosubstituted and disubstituted derivatives of acetone (CH_3COCH_2R and CH_3COCHR_2) or 1,3-diketones (CH_3COCH_2COR).

2. With stronger alkali, the molecule is attacked at both carbonyl groups to produce acetic acid and alkylated derivatives of that substance, as in the following examples:

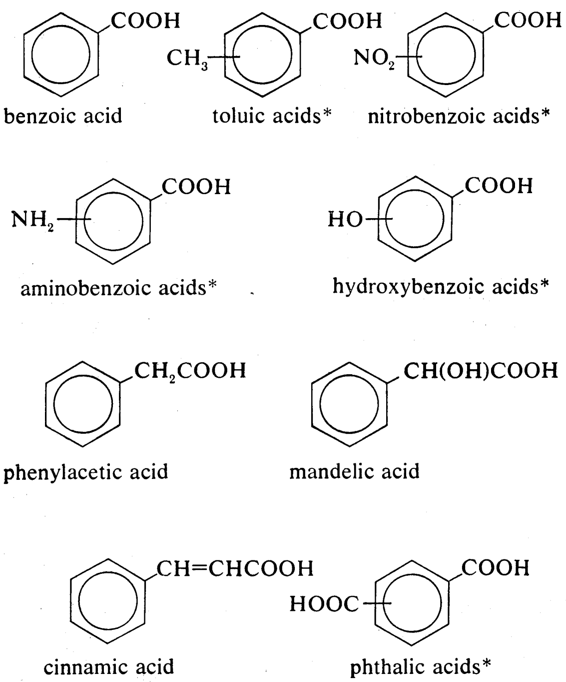


The polyfunctional nature of ethyl acetoacetate and related β-keto esters makes them valuable intermediates in the preparation of many heterocyclic compounds.

Reaction of aromatic carboxylic acids

Aromatic acids. In addition to the typical reactions undergone by the carboxyl group, aromatic acids also enter into the electrophilic substitution reactions commonly displayed by aromatic compounds. A carboxyl group directly attached to an aromatic ring exerts a deactivating influence and directs substituents to one of the two *meta* positions in the ring (*i.e.,* not to the adjacent carbon but the one next to that). Nitration of benzoic acid, for example, requires moderately vigorous conditions and gives first *m*-nitrobenzoic acid and then 3,5-dinitrobenzoic acid.

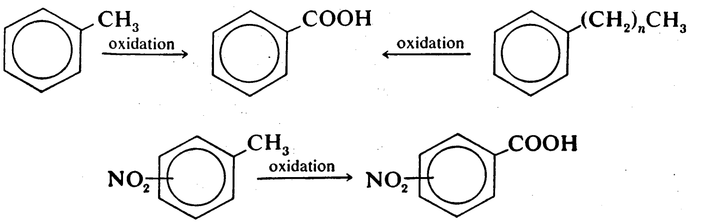
The simplest and best known aromatic acid is benzoic acid; this compound may be considered to be the parent of a wide range of substituted benzoic acids. Those acids in which the carboxyl group is part of an aliphatic side chain and not directly attached to the aromatic ring, such as phenylacetic, mandelic, and cinnamic acids, often are not classed as aromatic acids. Many polybasic acids, such as the isomeric phthalic acids, are commercially important. The most important aromatic acids are:



benzoic acid
toluic acids*
nitrobenzoic acids*
aminobenzoic acids*
hydroxybenzoic acids*
phenylacetic acid
mandelic acid
cinnamic acid
phthalic acids*

*These are convenient representations of the three isomeric acids (*ortho-*, *meta-*, and *para-*), which differ in the relative positions of attachment of the two groups on the ring.

Aromatic carboxylic acids, like their aliphatic counterparts, can be prepared by hydrolysis of nitriles or by the addition of carbon dioxide to organometallic compounds, usually derived from the appropriate halide; also, they often are obtained from aromatic hydrocarbons by oxidation. The aromatic ring is so resistant to oxidation that it remains intact while carbon-linked side chains are degraded to carboxyl groups. Certain noncarbon constituents such as nitro and halo groups survive this treatment, so that nitrotoluenes or halotoluenes may be oxidized to nitrobenzoic or halobenzoic acids:



Benzoic acid

Benzoic acid has been known for centuries, having been first obtained from gum benzoin, a resin extracted from an Indonesian tree. The well-known aromatic hydrocarbon benzene can be prepared by decarboxylation of benzoic acid. By standard procedures benzoic acid can be converted to the usual range of acyl derivatives, including benzoyl chloride, benzoate esters, and benzamides. The acid also enters into several aromatic substitution processes, although certain of the substituted benzoic acids are prepared preferably by alternative routes. Several of these substituted acids, such as salicylic acid (*o*-hydroxybenzoic) and anthranilic acid (*o*-aminobenzoic), are of considerable commercial importance. The latter, which is used as an intermediate in the preparation of many dyes, is generally prepared from phthalimide, itself a derivative of the dibasic acid phthalic acid. Salicylic acid is manufactured chiefly by the reaction of sodium phenolate (C_6H_5ONa) with carbon dioxide under pressure at 120°–140° C. Derivatives of salicylic acid, especially phenyl salicylate, acetylsalicylic acid (aspirin), and 4-aminosalicylic acid are used medicinally.

Phthalic acids

The phthalic acids are prepared by oxidation of appropriate aromatic hydrocarbons. The *ortho* isomer (in which the substituents are on adjacent positions on the ring) readily forms an anhydride. The *para* isomer (in which the substituents are located directly across from one another on the ring), which is also known as terephthalic acid, is used with dihydric alcohols to produce polymers, such as Terylene, which is a linear polymer that is made by the condensation of terephthalic acid and ethylene glycol.

Cinnamic acid, an important acid in which the carboxylic group is not directly attached to the aromatic ring, is prepared from benzaldehyde by several condensation procedures, one of which is carried out as follows:

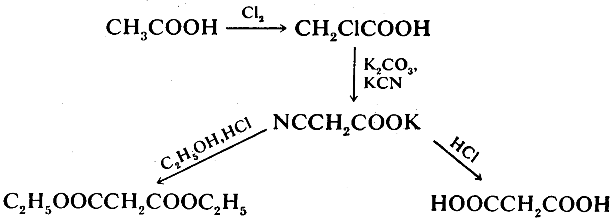
$$C_6H_5CHO \xrightarrow[CH_3COONa]{(CH_3CO)_2O} C_6H_5CH{=}CHCOOH.$$

Polycarboxylic acids. The most common dicarboxylic acids belong to the series of compounds with carboxyl groups at both ends of a hydrocarbon chain; that is, compounds with the general structural formula $HOOC(CH_2)_n$-COOH, in which n can be any of many numbers. Though they can be named systematically, several are generally designated by their common names, including oxalic ($n = 0$), malonic ($n = 1$), succinic ($n = 2$), glutaric ($n = 3$), adipic ($n = 4$), azelaic ($n = 7$), and sebacic ($n = 8$) acids.

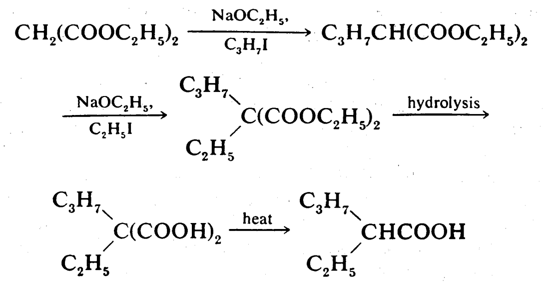
Oxalic acid, a common natural product, is readily obtained by oxidation of table sugar, or other carbohydrates, or by heating sodium formate. Oxalyl chloride (ClCO-COCl) is a useful reagent for preparing acyl halides from other carboxylic acids.

Malonic acid

Malonic acid and its more important ethyl ester are usually prepared from acetic acid by the following synthetic sequences:



The importance of diethyl malonate is that, like ethyl acetoacetate, it readily forms a resonance-stabilized carbanion, which readily can be alkylated and, by way of the corresponding free acid, decarboxylated. Diethyl malonate, therefore, can be used to prepare monosubstituted and disubstituted derivatives of acetic acid, as in the following example:

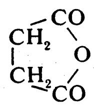


$$CH_2(COOC_2H_5)_2 \xrightarrow[C_3H_7I]{NaOC_2H_5,} C_3H_7CH(COOC_2H_5)_2$$

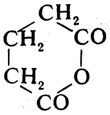
$$\xrightarrow[C_2H_5I]{NaOC_2H_5,} (C_3H_7)(C_2H_5)C(COOC_2H_5)_2 \xrightarrow{\text{hydrolysis}}$$

$$(C_3H_7)(C_2H_5)C(COOH)_2 \xrightarrow{\text{heat}} (C_3H_7)(C_2H_5)CHCOOH$$

Succinic and glutaric acids differ from the other members of this series in that they readily give cyclic anhydrides (see below *Anhydrides*) when heated. Oxalic and malonic acids fail to form cyclic anhydrides, and the members of the series above glutaric usually give linear polymeric anhydrides:

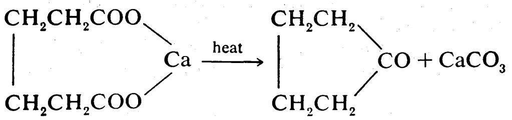


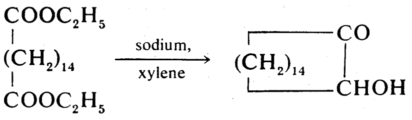
succinic anhydride



glutaric anhydride

Dibasic acids with six or more carbon atoms undergo cyclization reactions of various kinds. The examples given below show the production of a cyclic ketone, cyclopentanone, from the calcium salt of adipic acid and of an α-hydroxyketone (or acyloin) from a 16-carbon dicarboxylic acid ester:





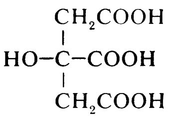
Adipic acid is used in the production of the most common form of nylon—a polyamide made from that dibasic acid and 1,6-hexanediamine. Maleic and fumaric acids are isomeric unsaturated dibasic acids. Malic acid, tartaric acid, and citric acid are important hydroxy-polycarboxylic acids of natural origin:

$HOOCCH(OH)CH_2COOH$

malic acid

$HOOCCH(OH)CH(OH)COOH$

tartaric acid



$$\begin{array}{c} CH_2COOH \\ | \\ HO{-}C{-}COOH \\ | \\ CH_2COOH \end{array}$$

citric acid

THE CARBOXYLIC ACID DERIVATIVES

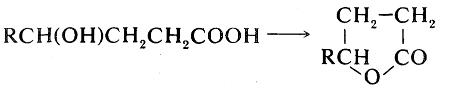
Esters. Formally, an ester with the structure RCOOR′ may be thought to be derived from a carboxylic acid of structure RCOOH, by replacement of hydrogen, —H, by an alkyl group, —R′, but the nature of the esterification process is such that it is more appropriate to describe the change as the replacement of the hydroxyl group, —OH, by an alkoxy group, —OR′. It is established experimentally in most esterification procedures that the alkoxy oxygen atom in an ester does, in fact, come from the alcohol—that is, that it remains attached to the R′ group.

Esterification

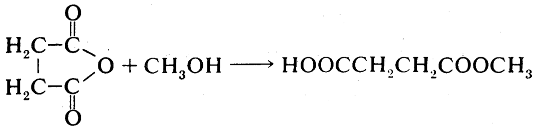
Esters are prepared most often by interaction of an acid and an alcohol in the presence of a catalyst, but other useful procedures also are available. The reversible reaction between alcohols and carboxylic acids is shown by the following equation:

$$RCOOH + R'OH \rightleftharpoons RCOOR' + H_2O.$$

Because the equilibrium is attained only slowly, it is usual to add a catalyst, such as hydrogen chloride or sulfuric acid, and to arrange the experimental conditions so that the equilibrium is shifted to the right. This shift is generally achieved by removal of water by distillation or by adding a large excess of one of the reactants. Hydroxy acids in which the hydroxyl group is three or four carbons from the carboxyl group undergo intramolecular reaction to give cyclic esters known as lactones:



Since acyl halides and acid anhydrides are more reactive than carboxylic acids, they can be esterified without the use of acidic catalysts. This method is especially useful for preparing the half esters of dibasic acids from their cyclic anhydrides and for producing the esters of phenols, which cannot be obtained directly from the acids:



succinic anhydride — methyl hydrogen succinate

$$CH_3COCl + C_6H_5OH \longrightarrow CH_3COOC_6H_5 + HCl.$$

acetyl chloride — phenol — phenyl acetate

The salts of carboxylic acids react with alkyl halides to form esters. The silver salts are most commonly employed for this purpose, but sodium salts may be used when the alkyl halide is a reactive one, *p*-nitrobenzyl bromide, for example:

$$p\text{-}NO_2C_6H_4CH_2Br + NaOOCCH_3 \longrightarrow p\text{-}NO_2C_6H_4CH_2OCOCH_3.$$

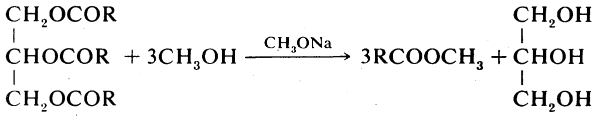
p-nitrobenzyl acetate

Methyl esters are conveniently obtained by reaction of carboxylic acids with diazomethane in ether solution:

$$RCOOH + CH_2N_2 \longrightarrow RCOOCH_3 + N_2.$$

Since esterification is a catalytic, reversible process, one ester can be prepared from another by reaction with an alcohol (alcoholysis), with an acid (acidolysis), or with a second ester (transesterification) in the presence of a suitable acidic or basic catalyst:

alcoholysis



$$\begin{array}{l} CH_2OCOR \\ | \\ CHOCOR \\ | \\ CH_2OCOR \end{array} + 3CH_3OH \xrightarrow{CH_3ONa} 3RCOOCH_3 + \begin{array}{l} CH_2OH \\ | \\ CHOH \\ | \\ CH_2OH \end{array}$$

acidolysis

$$RCOOR' + (NO_2)_2C_6H_3COOH$$

3,5-dinitrobenzoic acid

$$\xrightarrow{H_2SO_4} RCOOH + (NO_2)_2C_6H_3COOR'$$

transesterification

$$RCOOCH_3 + CH_3COOCH{=}CH_2$$

vinyl acetate

$$\xrightarrow{H_2SO_4} RCOOCH{=}CH_2 + CH_3COOCH_3$$

Esters are neutral compounds. The esters of long-chain acids occur naturally as waxes (esters of long-chain alcohols), as glycerides or fats (esters of glycerol), and as phosphoglycerides (esters of glycerophosphoric acid).

Reactions of esters

Esters can be converted to amides by reaction with ammonia, to hydrazides by reaction with hydrazine, and to acids by hydrolysis. They are not converted easily to acid chlorides or anhydrides:

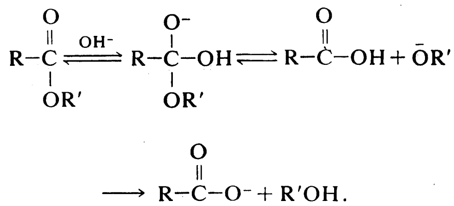
$$RCOOCH_3 + NH_3 \longrightarrow RCONH_2 + CH_3OH$$

$$RCOOCH_3 + N_2H_4 \longrightarrow RCONHNH_2 + CH_3OH$$

$$RCOOCH_3 + NaOH \longrightarrow RCOONa + CH_3OH.$$

Hydrolysis, the most important reaction, is brought about slowly by water and more quickly in the presence of acid or alkali. The alkaline hydrolysis, or saponification, of fats produces salts of long-chain acids, which are known as soaps. In common with many other long-chain compounds, these substances show surface-active properties and have long been used as cleansing agents. The enzyme-catalyzed hydrolysis of esters is important in the metabolism of lipids.

The mechanism of alkaline hydrolysis of esters is as follows:

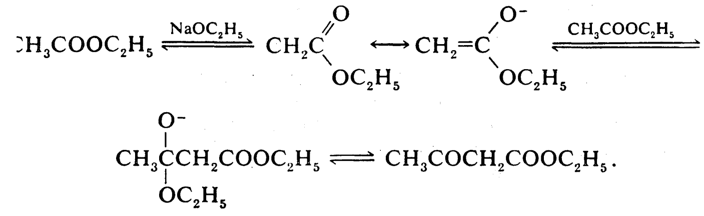


$$R-\underset{OR'}{\overset{O}{\overset{\|}{C}}} \xrightleftharpoons{OH^-} R-\underset{OR'}{\overset{O^-}{C}}-OH \rightleftharpoons R-\overset{O}{\overset{\|}{C}}-OH+\bar{O}R'$$

$$\longrightarrow R-\overset{O}{\overset{\|}{C}}-O^- + R'OH.$$

With sodium in an inert solvent, esters give acyloins (*α*-hydroxy ketones). Esters can be reduced to primary alcohols (RCH_2OH) by several methods, either catalytically with hydrogen and copper oxide–chromium oxide (300° C [572° F] and 250 atmospheres pressure), or chemically by reaction with sodium and alcohol or lithium aluminum hydride or diborane.

As is true of other carbonyl compounds, most esters undergo self-condensation in the presence of a base. The reaction involves the formation of a resonance-stabilized carbanion and reaction of this material with a second molecule of ester to give, finally, a *β*-keto ester. The best known example of this reaction is the preparation of ethyl acetoacetate from ethyl acetate:



$$CH_3COOC_2H_5 \xrightleftharpoons{NaOC_2H_5} CH_2C\begin{smallmatrix}\diagup O\\ \diagdown OC_2H_5\end{smallmatrix} \longleftrightarrow CH_2=C\begin{smallmatrix}\diagup O^-\\ \diagdown OC_2H_5\end{smallmatrix} \xrightleftharpoons{CH_3COOC_2H_5}$$

$$CH_3\underset{OC_2H_5}{\overset{O^-}{C}}CH_2COOC_2H_5 \rightleftharpoons CH_3COCH_2COOC_2H_5.$$

ethyl acetoacetate

Most *β*-keto esters contain a reactive methylene group that readily can be alkylated or acylated.

Anhydrides. Although thermal dehydration of acids is not usually a satisfactory process, anhydrides can be made from carboxylic acids by reaction with acetic anhydride, ketene, methoxyacetylene or ethoxyacetylene, or isopropenyl acetate:

$$2RCOOH+2CH_2=C=O$$

ketene

$$\longrightarrow 2RCOOCOCH_3 \longrightarrow (RCO)_2O+(CH_3CO)_2O$$

mixed anhydride

$$2RCOOH+HC\equiv COR' \longrightarrow (RCO)_2O+CH_3COOR'.$$

From acyl halides, anhydrides can be produced by reaction with the appropriate acid salt ($RCOCl + RCOONa \rightarrow (RCO)_2O + NaCl$). Acyl halides also react with carboxylic acid and pyridine or with acetic anhydride to produce anhydrides.

cetic
nhydride

Acetic anhydride, $(CH_3CO)_2O$, is of sufficient commercial importance to be prepared industrially by atmospheric oxidation of acetaldehyde in the presence of a metal acetate, as well as from acetic acid by reaction with acetylene or ketene:

$$CH_3CHO \xrightarrow{O_2} CH_3COOOH$$
$$\xrightarrow{CH_3CHO} (CH_3CO)_2O + H_2O$$

$$2CH_3COOH + HC\equiv CH \longrightarrow CH_3CH(OCOCH_3)_2$$
$$\longrightarrow (CH_3CO)_2O + CH_3CHO$$

$$CH_3COOH \xrightarrow{700°-800°} CH_2=C=O$$
$$\xrightarrow{CH_3COOH} (CH_3CO)_2O.$$

The anhydrides generally are useful acylating agents, being more reactive for this purpose than acids or esters, but less so than acyl halides. They react with water to give carboxylic acids, with alcohols or phenols to give esters, and with ammonia and amines to give amides. Acetic anhydride, a widely used acetylating agent, for example, reacts with water, an alcohol, and ammonia, according to the following equations:

$$(CH_3CO)_2O+H_2O \longrightarrow 2CH_3COOH$$
$$(CH_3CO)_2O+ROH \longrightarrow CH_3COOH+CH_3COOR$$
$$(CH_3CO)_2O+NH_3 \longrightarrow CH_3COOH+CH_3CONH_2.$$

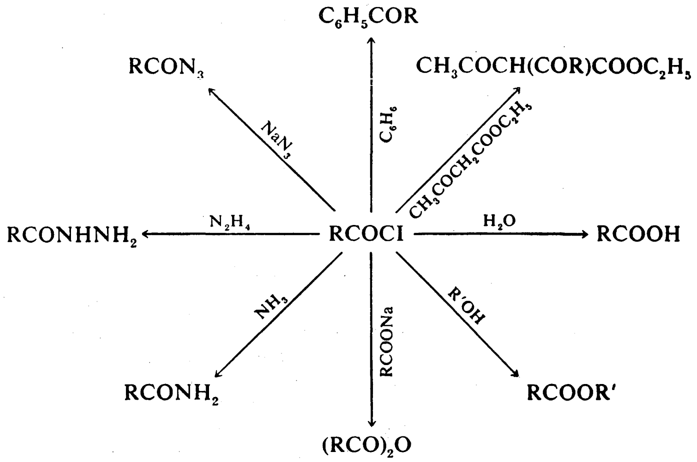
Anhydrides can also be used to acylate aromatic compounds in the presence of aluminum chloride.

Halides. The acyl halides, of which the chlorides are the best known, are among the most reactive of acyl compounds. They are easily made from carboxylic acids or their salts by reaction with phosphorus pentachloride, phosphorus trichloride, thionyl chloride (usually in the presence of pyridine), or oxalyl chloride. The reaction with thionyl chloride may be taken as typical:

$$RCOOH+SOCl_2 \longrightarrow RCOCl+SO_2+HCl.$$

Reactions
of acyl
halides

Acyl halides generally are vigorous acylating agents and react with water to give carboxylic acids; with alcohols and phenols to give esters; with sodium salts of carboxylic acids to give anhydrides; with ammonia and amines to give amides; with hydrazine to give hydrazides; and with sodium azide to give azides. They are also employed, with aluminum chloride as a catalyst, for the acylation of aromatic compounds; they interact with carbanions, such as that derived from ethyl acetoacetate, to give acyl derivatives. A summary of these reactions is shown below:



Acyl chlorides furnish ketones with organometallic compounds, especially those of zinc and of cadmium. They also are reduced to (1) aldehydes, by reaction with hydrogen, using a catalyst; or (2) alcohols, with lithium aluminum hydride or sodium borohydride. Aldehydes also can be prepared from acyl halides by way of *N,N*-dialkylamides, which are reduced with lithium aluminum hydride only to the aldehyde stage:

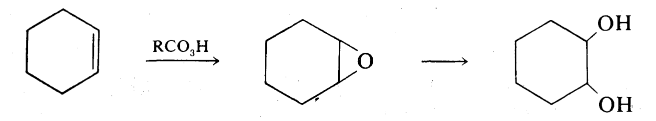
$$\underset{\text{acyl chloride}}{RCOCl} \xrightarrow{R'_2NH} \underset{N,N\text{-dialkylamide}}{RCONR'_2} \xrightarrow{LiAlH_4} \underset{\text{aldehyde}}{RCHO.}$$

Peroxy acids. The peroxy acids, or peracids as they are sometimes called, may be considered as monoacyl derivatives of hydrogen peroxide. They are important oxidizing agents and are prepared from carboxylic acids and hydrogen peroxide in the presence of acidic catalysts (usually sulfuric acid or methanesulfonic acid) or from acid anhydrides and hydrogen peroxide. It is not always necessary to isolate the peracid, and mixtures of carboxylic acid and hydrogen peroxide may be employed as oxidizing agents.

Peracetic acid is manufactured by the atmospheric oxidation of acetaldehyde.

Reactions of peracids

As oxidizing agents, the peracids are used (1) to prepare epoxides and 1,2-diols from alkenes; (2) to convert ketones to esters; and (3) to oxidize amines to amine oxides, nitroso compounds, or nitro compounds. Epoxides result when alkenes are treated with aromatic peracids, such as peroxybenzoic or monoperoxyphthalic acid, or with aliphatic peracids in a buffered solution. Without a buffer the reaction mixture containing the aliphatic acid is so acidic that the epoxide ring is cleaved to give the 1,2-diol:



cyclohexene epoxycyclohexane cyclohexane-1,2-diol

The aliphatic peracids most widely used are peroxyformic, peroxyacetic, and peroxytrifluoroacetic.

Nitrogen-containing derivatives. *Nitriles.* Nitriles, or cyanides, with the general structure RC≡N, are made (1) from alkyl halides (or sulfates) by reaction with sodium or potassium cyanide, (2) from acids by dehydration of the corresponding amides, or (3) from aldehydes by dehydration of the aldoximes. These reactions are shown below:

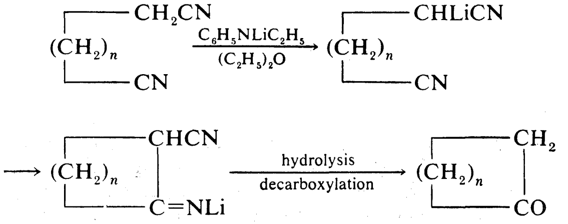
$$RX \longrightarrow RCN$$
$$RCOOH \longrightarrow RCONH_2 \longrightarrow RCN$$
$$RCHO \longrightarrow RCH{=}NOH \longrightarrow RCN.$$

Reactions of nitriles

Nitriles can be hydrolyzed to amides and to acids, or they can be converted directly to esters. They are reduced easily to primary amines; but reaction occurs via imines and, under suitable conditions, reduction can be halted at this stage, with the imine being hydrolyzed subsequently to an aldehyde:

$$RCN \xrightarrow{\text{hydrolysis}} RCONH_2 \xrightarrow{\text{hydrolysis}} RCOOH$$
$$RCN \xrightarrow[(C_2H_5)_2O]{SnCl_2,HCl,} RCH{=}NH \xrightarrow{H_2O} RCHO.$$

Nitriles share with esters the ability to undergo base-catalyzed self-condensation, one molecule acting as a nitrile and the second as a source of carbanion. Dinitriles undergo internal condensation and are used to prepare cyclic compounds in this way:


$$(CH_2)_n\begin{cases} CH_2CN \\ CN \end{cases} \xrightarrow[(C_2H_5)_2O]{C_6H_5NLiC_2H_5} (CH_2)_n\begin{cases} CHLiCN \\ CN \end{cases}$$
$$\longrightarrow (CH_2)_n\begin{cases} CHCN \\ C{=}NLi \end{cases} \xrightarrow[\text{decarboxylation}]{\text{hydrolysis}} (CH_2)_n\begin{cases} CH_2 \\ CO \end{cases}$$

Amides. The preparation of amides from acyl halides or anhydrides (or sometimes esters) is generally accomplished by reaction with ammonia or with a primary or secondary amine. Amides also are the products of the partial hydrolysis of nitriles, and they can be obtained directly from carboxylic acids by reaction with urea, sulfamide (in the presence of pyridine), thiourea, or hexamethylphosphoramide.

Reactions of amides

Amides are hydrolyzed by acid or alkali to carboxylic acids, and they are dehydrated readily to nitriles. Amides are normally reduced to amines, but *N,N*-dialkylated amides give aldehydes. With bromine and alkali, amides are degraded to amines with one less carbon atom, as follows:

$$RCONH_2 \xrightarrow{Br_2,NaOH} [RCONHBr] \longrightarrow [RNCO] \longrightarrow RNH_2 + CO_2.$$

Urea (NH_2CONH_2) is an amide present in mammalian urine and is the form in which many animals excrete waste nitrogen.

Hydrazides and azides. Hydrazides ($RCONHNH_2$) may be considered formally to be derived from carboxylic acids and hydrazine, and acyl azides ($RCON_3$) are analogous to the acyl chlorides, the azide (N_3) group having replaced a chlorine (Cl) atom. In actual practice, the hydrazides and azides are prepared from acyl halides by reaction with hydrazine and sodium azide, respectively. Nitrous acid converts hydrazides to azides:

$$\begin{tikzcd} & RCONHNH_2 \arrow[dd, "HNO_2"] \\ RCOCl \arrow[ur, "N_2H_4"] \arrow[dr, "NaN_3"'] & \\ & RCON_3 \end{tikzcd}$$

THE IDENTIFICATION AND ANALYSIS OF CARBOXYLIC ACIDS AND DERIVATIVES

In the qualitative analysis of organic compounds, carboxylic acids are recognized by their solubility in aqueous alkali and by their ability to liberate carbon dioxide from bicarbonate solutions. When treated with potassium hydroxide and hydroxylamine, esters give hydroxamic acids, which are easily recognized by the deep red or purple colours they give with ferric salts.

Infrared spectra of carboxyl compounds

All acyl compounds have distinctive bands in their infrared spectra (particular patterns of light absorption, instrumentally determined). The absorption is believed to result from energy taken up by certain carbonyl stretching motions in the molecules. The exact frequency of this absorption depends on the molecular environment of the carbonyl group, and it is sometimes possible to distinguish the various types of carbonyls by the frequency of the infrared absorption (see Table 25). A double bond, or an aromatic system in the alpha position with reference to the carbonyl group, causes the absorption in each case to appear at the lower end of the range.

Table 25: Characteristic Infrared Absorption Bands of Selected Classes of Carbonyl Compounds

class	infrared absorption frequency, C=O stretching (cm^{-1})
Anhydrides	1850–1740*
Acyl halides	1815–1750
Esters	1750–1710
Aldehydes	1740–1680
Ketones	1725–1660
Carboxylic acids	1720–1660
Amides	1680–1630

*Two peaks, about 60 cm^{-1} apart.

Carboxylic acids can be characterized by the preparation of a wide range of solid derivatives. Among the most commonly employed are the

amides	($RCONH_2$),
anilides	($RCONHC_6H_5$),
p-toluidides	($RCONHC_6H_4CH_3$),
p-nitrobenzyl esters	($RCOOCH_2C_6H_4NO_2$),
S-benzylthiuronium salts	[$RCO\overset{-}{O}(\overset{+}{N}H_2)_2CSCH_2C_6H_5$],
p-bromophenacyl esters	($RCOOCH_2COC_6H_4Br$),
and hydrazides	($RCONHNH_2$).

Esters and other acyl derivatives usually furnish these compounds after hydrolysis to the acid; or, in some cases, they are converted directly to the compound without production of the free acid.

The reaction of carboxylic acids and esters with alkali is the basis of several procedures for the quantitative estimation of these groups. Acids react quickly at room temperature and can be titrated directly in aqueous or alcoholic solution. Alternatively, they can be treated with excess of alkali, the excess being titrated with standard acid solution. Esters usually react with hot alkali, and the normal procedure for their quantitative determination is heating with excess alkali and titration of the surplus.

This procedure is the basis of the standard method for determining the saponification (hydrolysis) equivalent, or saponification value, of fats. Carboxylic acids can also be quantitatively estimated by determination of (1) the methane liberated during reaction with methylmagnesium iodide or (2) the water formed during esterification.

hroma-
ography of
arboxyl
om-
ounds

Chromatography is a method of separation based on differential adsorption of substances from solution by a solid material (such as alumina). Most of the known chromatographic procedures have been used to separate mixtures of carboxylic acids, as well as mixtures of the various derivatives, and many of the methods have been adapted for qualitative or quantitative identification or both. Indeed many useful developments in chromatography have come about from the need to analyze the mixed amino acids resulting from protein hydrolysis and the mixed fatty acids produced by lipid hydrolysis.

The most important analytical procedure for esters is the form of chromatography known as gas–liquid chromatography. By the use of an appropriate stationary phase, and long capillary columns when necessary, even closely related esters can be separated and quantitatively analyzed. In difficult cases this technique may have to be allied with other separation procedures.

Limited information about the carboxyl group (and other acyl groups) is obtainable by nuclear magnetic resonance spectroscopy, an instrumental technique that gives information about the magnetic properties of atomic nuclei. Although this method may not reveal much about the carboxyl group, it may provide considerable useful information about the remainder of the molecule. Mass spectrometry, the separation of substances on the basis of their respective masses, is an increasingly important tool in the recognition of complex compounds containing carboxyl and other acyl groups, especially those derived from lipids and other natural sources. Fragmentation of the molecule during the mass spectrometry procedure frequently gives valuable information about the molecular structure of the compound or compounds under investigation. Esters, particularly methyl esters, have been more extensively used in mass spectrographic studies than the acids themselves or the other derivatives. (F.D.G.)

Organic halogen compounds

Organic halogen compounds are substances containing atoms of one or more of the so-called halogen elements (fluorine, chlorine, bromine, and iodine) joined to atoms of carbon. In their physical properties and chemical behaviour they are much like other organic compounds (compounds of carbon). Very few organic halogen compounds occur in nature; most are products of the organic chemist's laboratory. Many organic halogen compounds are used as solvents; carbon tetrachloride, for example, is employed as a cleaning fluid. Other members of the family are used as anesthetics and refrigerants. The plastic material polyvinyl chloride is an organic halogen compound; so is the herbicide 2,4-D (2,4-dichlorophenoxyacetic acid). Many organic halogen compounds are used in industry or in the laboratory for conversion to other compounds with useful properties, such as dyes and medicinal agents.

GENERAL CONSIDERATIONS

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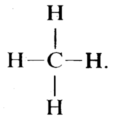
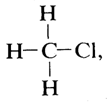
Chemically, organic halogen compounds may be considered to be derivatives of hydrocarbons (compounds composed exclusively of carbon and hydrogen), in which certain of the hydrogen atoms have been replaced by halogen atoms. In the molecules of hydrocarbons, the carbon atoms form a framework, or skeleton, to which the hydrogen atoms are attached. When the halogen atoms replace hydrogen atoms, they too occupy peripheral or terminal positions, leaving the central framework positions in the molecule to the carbon atoms.

In theory, successive replacement of hydrogen atoms with halogen atoms is possible in any hydrocarbon molecule, until all hydrogen atoms have been replaced. The simplest hydrocarbon, for example, is methane, the molecules of which are composed of one carbon atom and four hydrogen atoms. The structural formula of methane is



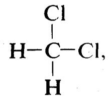
$$\begin{array}{c}\mathrm{H}\\ |\\ \mathrm{H{-}C{-}H.}\\ |\\ \mathrm{H}\end{array}$$

Replacement of one of the hydrogen atoms in methane with a chlorine atom (Cl) gives the substance methyl chloride



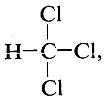
$$\begin{array}{c}\mathrm{H}\\ |\\ \mathrm{H{-}C{-}Cl,}\\ |\\ \mathrm{H}\end{array}$$

or CH_3Cl. Further replacements of hydrogen with chlorine give methylene dichloride



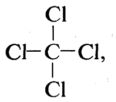
$$\begin{array}{c}\mathrm{Cl}\\ |\\ \mathrm{H{-}C{-}Cl,}\\ |\\ \mathrm{H}\end{array}$$

or CH_2Cl_2; chloroform



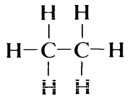
$$\begin{array}{c}\mathrm{Cl}\\ |\\ \mathrm{H{-}C{-}Cl,}\\ |\\ \mathrm{Cl}\end{array}$$

or $CHCl_3$; and carbon tetrachloride



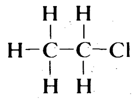
$$\begin{array}{c}\mathrm{Cl}\\ |\\ \mathrm{Cl{-}C{-}Cl,}\\ |\\ \mathrm{Cl}\end{array}$$

or CCl_4. Similar replacements can be carried out in molecules of the hydrocarbon ethane



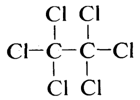
$$\begin{array}{c}\mathrm{H\ \ H}\\ |\ \ \ |\\ \mathrm{H{-}C{-}C{-}H}\\ |\ \ \ |\\ \mathrm{H\ \ H}\end{array}$$

(C_2H_6) to give a series of compounds ranging from ethyl chloride



$$\begin{array}{c}\mathrm{H\ \ H}\\ |\ \ \ |\\ \mathrm{H{-}C{-}C{-}Cl}\\ |\ \ \ |\\ \mathrm{H\ \ H}\end{array}$$

(C_2H_5Cl) to hexachloroethane (C_2Cl_6).



$$\begin{array}{c}\mathrm{Cl\ \ Cl}\\ |\ \ \ \ |\\ \mathrm{Cl{-}C{-}C{-}Cl}\\ |\ \ \ \ |\\ \mathrm{Cl\ \ Cl}\end{array}$$

Atoms of each of the other halogen elements—fluorine (F), bromine (Br), and iodine (I)—can similarly be substituted for hydrogen atoms in hydrocarbons (giving compounds containing atoms of more than one halogen element, as well as compounds with multiple atoms of any one element). When the precise nature of the halogen atom is not important, it is customary to use the symbol X, which can stand for an atom of any halogen element. Thus CH_3X could mean methyl fluoride (CH_3F), methyl chloride (CH_3Cl), methyl bromide (CH_3Br), or methyl iodine (CH_3I).

Although organic halogen compounds can be considered to be derived formally from hydrocarbons by replacement of hydrogen atoms with halogen atoms, this often is not the procedure by which organic halogen compounds are actually prepared in the laboratory. As will be shown, a variety of synthetic procedures is available for preparing organic halogen compounds from other classes of organic compounds.

Nomenclature and classification. Organic halogen compounds, generally, are named according to one of two principles, substitution or the use of organic radicals. In substitution names, the combining form of the name of the halogen element is used as a prefix to the name of the parent hydrocarbon. In this system the compound CH_3Cl is called chloromethane. The analogous combining forms of the other halogen elements are fluoro-, bromo-, and iodo-; halo- is the combining form of the generic term halogen.

In the "radical" system of nomenclature the organic portion of the molecule (everything but the halogen atom) is designated by a combining form, or radical, which is followed by the name of the halogen in its -ide form (*e.g.*, chloride for chlorine). The organic radicals are derived from the names of the hydrocarbons by substituting the ending -yl for a portion of the hydrocarbon name. Thus methane becomes methyl, and the substance designated chloromethane above is also termed methyl chloride. Often, when it is desired not to specify one particular radical, the letter R— is used. In this way $R-Cl$ represents any organic chlorine compound, and the designation $R-X$ stands for the entire class of organic monohalogen compounds.

Almost any class of organic chemical compound—including alcohols, aldehydes, ketones, and carboxylic acids—can carry halogen atoms in its molecules. The phrase "organic halogen compounds," however, is usually taken to refer simply to halogen derivatives of the various hydrocarbons.

General chemical properties. As previously noted, the chemical behaviour of chemical compounds is largely determined by the electronic configuration of their molecules, and this in turn depends upon the electronic configurations of the constituent atoms. The electronic configuration of halogen atoms is quite different from those of carbon and hydrogen atoms, with the result that the incorporation of halogen atoms into hydrocarbon molecules has a marked effect on their chemical behaviour.

Effect of polarizability

Atoms contain varying numbers of negatively charged electrons in what may be defined crudely as concentric shells about the positively charged nucleus. The atoms of the halogen elements, as a group, are characterized by the fact that their outermost electron shells include seven electrons. The halogen elements are differentiated from one another by the number of electron shells, and consequently by the total number of electrons they contain. These numbers increase with the atomic weights of the halogen elements, in the following order: fluorine, chlorine, bromine, and iodine. The seven electrons in the outermost shells, the so-called valence electrons, are farther and farther from the nucleus as the number of shells increases (that is, in the order given above). (The atomic radius of the iodine atom, for example, is rather more than twice that of the fluorine atom.) Consequently the valence electrons are most firmly held by the fluorine nucleus and least firmly held by the iodine nucleus. Because of its small size, fluorine is the most electronegative (electron-attracting) of the halogens, and iodine is the least. In general, less strongly held electrons, such as those of iodine, are more polarizable; *i.e.*, more easily distorted by attractive or repulsive forces near them. Polarizability is an important property that greatly influences the behaviour of halogen atoms with regard to organic molecules, as for example, when free halide ions (halide atoms containing an extra electron, and hence a net negative charge) attack positive centres in organic compounds (as occurs in the synthesis of many organic halogen compounds). In such "nucleophilic" (or positive-charge-seeking) attack, the halide ions attack preferentially in the decreasing order: iodide, bromide, chloride, and fluoride. This explains why hydriodic acid reacts more readily with alcohols to give alkyl halides than do the other halogen acids.

Because carbon is less electronegative than any of the halogen elements, the halogen atom attracts the two electrons that comprise a carbon–halogen bond more strongly than does the carbon atom. As a result, the electrons of the bond are displaced toward the halogen atom, and the bond is polarized. The carbon atom concerned will thus bear a partial positive charge, and will be susceptible to attack by nucleophiles—reagents that seek positive centres. Typical nucleophiles are negatively charged groups, such as the hydroxide ion, a negatively charged group comprised of a linked hydrogen and oxygen atom, OH^-, and molecules having atoms with lone pairs of electrons, such as ammonia, $:NH_3$ (in which the two dots before the N represent a lone pair of electrons). The result of nucleophilic attack on an organic halide is substitution (or replacement) of the halide by the nucleophile. The equation below represents a nucleophilic substitution reaction in which the iodine atom in methyl iodide is replaced by the hydroxide ion acting as a nucleophile:

$$HO^- + \overset{\delta+}{C}H_3 - \overset{\delta-}{I} \longrightarrow HO - CH_3 + I^-$$

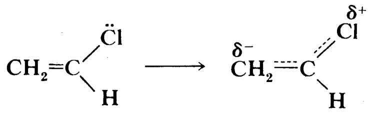
As is usually the case in chemical equations the starting materials are written at the left of the arrow and the products of the reaction are written at the right. In this equation, the polarization of the methyl iodide molecule is indicated by the symbols δ^+ and δ^-, in which δ (delta) means that partial charges (with signs as shown) appear on the atoms in question.

The readiness with which such nucleophilic substitutions occur depends on two factors: (1) the ease of dissociation of the carbon–halogen bond and (2) the nature of the leaving group (the halide ion). The ease of dissociation of a bond is inversely proportional to the bond energy, which for a carbon–halogen bond increases in the order: iodine, bromine, chlorine, and fluorine. A good leaving group, in general, is the anion of a strong acid, and the strengths of the halogen acids decrease in the order hydrogen iodide, hydrogen bromide, hydrogen chloride, and hydrogen fluoride. The bond dissociation energies and the nature of the leaving groups thus act in the same direction for the halides, with the result that the susceptibility of the various alkyl halides to nucleophilic attack decreases in the order: iodide, bromide, chloride, and fluoride.

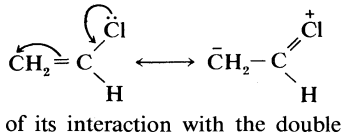
Free-radical reactions

Organic halogen compounds also undergo free-radical reactions, reactions that lead to splitting of a bond, one electron going with each fragment. In such reactions, only the bond dissociation energies are significant, and, in this case, the order of reactivity of the alkyl halides is the same as it is in nucleophilic attack.

When a halogen atom is attached to an unsaturated carbon atom (one in which not all the valence bonds are saturated), as in certain alkenyl halides, such as vinyl chloride, $CH_2{=}CHCl$, the halogen atom still attracts electrons to it, because of its electronegativity, but another factor must also be considered. In vinyl chloride, for example, the orbital (position in space) of one of the lone pairs of electrons of the chlorine atom overlaps with the orbital of certain of the electrons associated with the double bond. These are the so-called pi electrons, two of which (one from each carbon atom) occupy an orbital covering both carbon atoms and form one of the bonds (called a pi bond) uniting the two carbon atoms. These pi bonds occur only in multiple bonds. As a result of the interaction of the chlorine electrons with the pi electrons of the double bond, carbon–chlorine bonds of this particular type themselves have some double-bond character. Since this involves a partial loss of electrons from chlorine and a gain by the double bond, the molecule becomes polarized as shown in the equation below:



in which the dotted lines indicate partial double bond character. This phenomenon also can be described in a different way as a movement of the lone pair of electrons associated with the halogen atom, as shown in the diagram below by a curved arrow. This shift of electrons causes a polarization of the double bond. The actual state of the molecule is between the two extreme forms shown below, connected by a double-headed arrow:

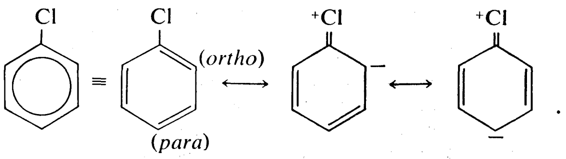


As a result of its interaction with the double bond, the halogen atom of a vinyl halide is much less susceptible to nucleophilic substitution than is an ordinary alkyl halide.

The presence of halogen atoms adjacent to a double bond—the situation in vinyl halides—also has an effect on the chemical reactivity of the double bond. Generally, one of the principal reactions of double bonds is electrophilic addition—a reaction that is initiated by an electron-seeking (electrophilic) reagent and leads to saturation of the double bonds, that is, to production of a molecule with nothing but single bonds. Electrophilic addition is strongly influenced by the polarization of the double bond caused by the halogen atom. As indicated above, a halogen atom on one carbon atom of a double bond leads to a partial negative charge on the other atom of the double bond. This negative charge is the target of an electrophilic reagent, so that the direction of electrophilic addition is dictated by the halogen atom.

When there are several halogen atoms attached to the double bond, their electron-attracting powers become more evident; in fact, they may reduce the negative character of the double bond to such an extent that it adds electrophilic reagents very much less readily than does a double bond without halogen substituents. At the same time, however, it may become more reactive toward nucleophilic reagents (which attack positive centres).

The same situation arises in the aryl halides. In this case, however, there is a more extensive unsaturated system, which can polarize in several different ways, as shown below:



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In these diagrams, the hexagon represents a ring of six carbons (carrying single hydrogen atoms except where the chlorine atom is shown); the aromatic nature of the ring system is indicated either by a circle within the hexagon or by three double bonds (the two forms being connected by an identity sign). Forms in which separation of charges occurs (as shown by the plus and minus signs) are connected by double-headed arrows, indicating that they contribute to the overall structure of the molecule. Because this polarization produces partial negative charges at the positions in the ring designated *ortho* and *para* (from Greek roots chosen somewhat arbitrarily before the correct positions were known), electrophilic substitution (*i.e.*, attack by positively charged groups) occurs preferentially at these positions. In aromatic compounds of most varieties, substitution reactions occur more readily than addition reactions because addition destroys the aromaticity of the ring, whereas substitution does not. In polyhalogenated aromatic derivatives, the electronegative halogen atoms reduce the susceptibility of the ring to electrophilic substitution by withdrawing electrons generally from the ring, but, at the same time, they activate the molecule toward nucleophilic attack, of such nature that the halogen atoms themselves are replaced by the attacking nucleophiles.

General physical properties. Generally in organic halogen compounds, the boiling points increase as the atomic weights of the halogen elements increase and as the molecular weights of the compounds themselves increase (see Table 26). On the other hand, the more highly branched the chain of a given number of carbon atoms is, the lower the boiling point of the compound is.

Table 26: Boiling Points of Alkyl Halides (°C)

radical	formula	fluoride	chloride	bromide	iodide
Methyl	CH_3—	−78.4	−23.76	3.45	42.5
Ethyl	CH_3CH_2—	−38.0	+12.5	38.4	72.4
n-Propyl	$CH_3CH_2CH_2$—	− 2.5	46.6	71.0	102.5
Isopropyl	$(CH_3)_2CH$—	− 9.4	34.8	59.4	89.5
n-Butyl	$CH_3(CH_2)_3$—	32.5	78.5	101.5	130.4
Isobutyl	$(CH_3)_2CHCH_2$—	. . .	68.8	91.4	121.0
sec-Butyl	$C_2H_5CH(CH_3)$—	25.1	68.2	91.2	120.0
tert-Butyl	$(CH_3)_3C$—	12.1	50.7	73.2	103.3

Boiling points usually increase with the number of halogen atoms; *e.g.*, methyl chloride (CH_3Cl), −24° C (−11.2° F); methylene chloride (CH_2Cl_2), 40° C (104° F); chloroform ($CHCl_3$), 62° C (143.6° F); carbon tetrachloride (CCl_4), 77° C (170.6° F).

Densities increase with increase in the atomic weight of the halogen and also with the number of halogen atoms in the molecule; *e.g.*, ethyl fluoride (C_2H_5F), 0.818; ethyl chloride (C_2H_5Cl), 0.923; ethyl bromide (C_2H_5Br), 1.47; ethyl iodide (C_2H_5I), 1.98; ethylidene chloride (CH_3CHCl_2), 1.18; ethylene dichloride (CH_2ClCH_2Cl), 1.25; 1,1,1-trichloroethane CH_3CCl_3, 1.35; 1,1,2-trichloroethane $CH_2ClCHCl_2$, 1.44 (the numbers in the names designating the carbon atoms to which the halogens are attached).

Melting
and boiling
points

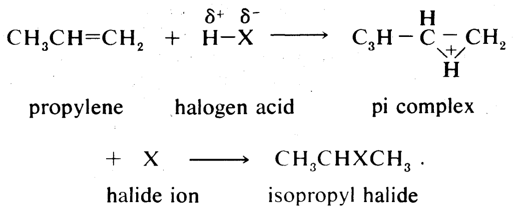
Melting points and boiling points of organic halogen compounds, however, do not depend exclusively on the molecular weights of the molecules concerned; they also are influenced by the intermolecular forces involved and hence on the polar (unsymmetrically charged) nature of the molecule (polar molecules being more attracted to one another than nonpolar molecules and hence more resistant to both melting and boiling). For instance, geometrical isomers—that is, compounds with the same structural framework but different spatial arrangements of the substituent groups of atoms—have different melting and boiling points (which also differ from those of any other isomers of the same compounds) because the halogen atoms are differently oriented with respect to one another. An example of such a circumstance is found in the difluoroethylenes: 1,1-difluoroethylene, $CH_2{=}CF_2$, boils at −84° C (−119.2° F); whereas the two geometrical isomers of 1,2-difluoroethylene, $CHF{=}CHF$, which are known as *cis*- and *trans*-1,2-difluoroethylene, boil at −26° C (−15° F) and −53° C (−43° F) respectively. There is a considerable variation in physical properties among polyhalogenated compounds also because of differences in polarity. For example, a nonpolar fully fluorinated compound often has a lower boiling point than less highly fluorinated compounds with the same carbon skeleton, chiefly because the latter are more polar (see Table 27). Boiling points of representative alkyl halides are given in Table 26, and melting and boiling points of aryl halides are presented in Table 27.

The weak intermolecular forces in fully fluorinated compounds (resulting from their low polarity) also are responsible for the low viscosities these substances exhibit—they are used as oils and greases under carefully controlled conditions. (Their viscosities decreased rapidly with increase in temperature, however, and this fact prevents their more general use.)

The bond dissociation energies of carbon-to-halogen bonds, given in the standard units of kilocalories per mole (a mole being equal to the molecular weight expressed in grams), are carbon–fluorine, 108; carbon–chlorine, 81; carbon–bromine, 67; and carbon–iodine, 53. For purposes of comparison, it may be noted that the carbon–carbon and carbon–hydrogen bond dissociation energies are 84 and 102 kilocalories per mole, respectively. The high dissociation energy of the carbon–fluorine bond and the low dissociation energy of the carbon–iodine bond help to explain why the fluorocarbons are thermally stable whereas organic iodine compounds dissociate readily on heating (with liberation of iodine).

ALKYL HALIDES

Monohalides. *Preparation.* Hydrocarbons bearing one halogen substituent are most commonly prepared by two methods: (1) addition of halogen acids to olefins (alkenes) and (2) replacement of hydroxyl (−OH) groups in alcohols. In its simplest form addition of halogen acids to olefins is an electrophilic addition reaction in which the positively charged end of the polarized halogen acid is attracted to the olefinic double bond. The acid then dissociates producing a halide ion and a proton (hydrogen ion) which, with the pi electrons of the double bond, forms a positively charged intermediate known as a pi complex. Finally, the halide ion reacts with the pi complex to give the ultimate product. A typical reaction of this type is shown in chemical formulas below:



$$\mathrm{CH_3CH{=}CH_2} + \overset{\delta^+\ \delta^-}{\mathrm{H{-}X}} \longrightarrow \mathrm{C_3H{-}\overset{H}{C}\underset{H}{\overset{+}{-}}CH_2}$$

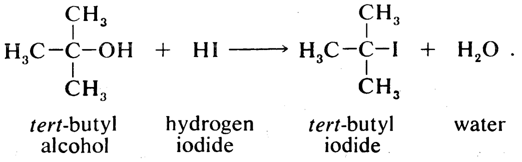
$$+\ \mathrm{X} \longrightarrow \mathrm{CH_3CHXCH_3}\,.$$

The direction of addition to the double bond can be predicted by a consideration of the effects of the alkyl groups (in this case, only methyl, CH_3-) attached to the double bond carbons. The various halogen acids undergo addition to double bonds, in decreasing order of reactivity, as follows: hydrogen iodide, hydrogen bromide, hydrogen chloride, and hydrogen fluoride.

In the presence of traces of peroxides (which can be produced from olefins and traces of oxygen by sunlight), hydrogen bromide reacts with olefins in a different fashion, often to give products in which the addition occurs in the opposite direction to that in the above reaction (that is, with bromine atom appearing on the other double bond carbon atom).

Conversion of alcohols to halides

Alkyl halides are produced from alcohols by means of nucleophilic substitution reactions with hydrogen halides (halogen acids), phosphorus halides, or thionyl halides. The reactivities of the halogen acids in this reaction are, in decreasing order: hydrogen iodide, hydrogen bromide, and hydrogen chloride. (Hydrogen fluoride cannot be used.) Tertiary alcohols (alcohols with three carbon atoms attached to the hydroxyl carbon) undergo this reaction readily. Secondary and primary alcohols (alcohols with two and one carbon atom joined to the hydroxyl carbon, respectively) react with less facility. For example, tertiary butyl alcohol reacts with aqueous hydrogen iodide at room temperature, according to the following equation:



$$\mathrm{H_3C{-}\underset{CH_3}{\overset{CH_3}{C}}{-}OH} + \mathrm{HI} \longrightarrow \mathrm{H_3C{-}\underset{CH_3}{\overset{CH_3}{C}}{-}I} + \mathrm{H_2O}\,.$$

Secondary bromides and iodides (corresponding in structure to secondary alcohols) usually can be prepared in the same way. For primary alkyl chlorides, however, it is necessary to use a catalyst and warm the solution. (Long heating may cause isomerization—rearrangement to another product.)

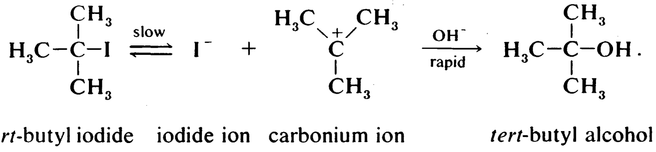
Phosphorus pentachloride, phosphorus tribromide, red phosphorus and bromine, phosphorus triiodide, and red phosphorus and iodine all react with the appropriate alcohols to give primary, secondary, and tertiary halides. Thionyl chloride and thionyl bromide react with primary or secondary alcohols in the presence of an organic base, such as pyridine. Tertiary alcohols need no base.

The various alkyl halides can be interconverted, and they also can be prepared from other alkyl compounds. Either process is one of nucleophilic substitution by halide ions, and the choice of solvent for the reaction is important.

Reactions. In most of their reactions, the alkyl iodides are more reactive than bromides, which are more reactive than the chlorides, and these, in turn, are more reactive than the fluorides. This order of reactivity reflects the relative ease with which the respective carbon–halogen bonds are broken. Also, generally speaking, the tertiary halides react most readily, the secondary less so, and the primary least of all, reflecting the fact that alkyl substituents tend to donate electrons to the carbon atom that carries the halogen, causing it to release the halide ion, and the more alkyl substituents there are (three in tertiary compounds), the more easily is the halide ion lost. The chemical reactions undergone by alkyl halides can be classified under three headings: (1) nucleophilic substitutions, (2) elimination reactions, and (3) free-radical processes.

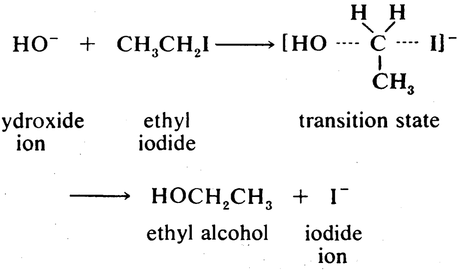
Nucleophilic substitutions

Nucleophilic substitution reactions are classified as unimolecular or bimolecular according to whether the reaction involves one or two molecules. Unimolecular nucleophilic substitutions proceed by way of ionization of the alkyl halide—that is, by cleavage of the molecule to give a positively charged alkyl, or carbonium, ion and a negatively charged halide ion. This type of reaction is undergone most readily by a tertiary alkyl halide in ionizing solvents (aqueous alcohol, for example). Reactions of this type are known to chemists as S_N1 reactions (meaning substitution, nucleophilic, unimolecular). A good example of an S_N1 reaction is the hydrolysis of *tert*-butyl iodide:



$$\mathrm{H_3C{-}\underset{CH_3}{\overset{CH_3}{C}}{-}I} \underset{}{\overset{\text{slow}}{\rightleftharpoons}} \mathrm{I^-} + \underset{\mathrm{CH_3}}{\overset{\mathrm{H_3C\ \ CH_3}}{\overset{+}{\mathrm{C}}}} \xrightarrow[\text{rapid}]{\mathrm{OH^-}} \mathrm{H_3C{-}\underset{CH_3}{\overset{CH_3}{C}}{-}OH}\,.$$

Bimolecular nucleophilic substitutions—S_N2 reactions—proceed through a transition state that includes a nucleophile—for example a hydroxide ion—as well as a molecule of the alkyl halide. They are most common for primary alkyl halides in nonionizing solvents. An example is the reaction of ethyl iodide with a hydroxide ion to give ethyl alcohol and the iodide ion:



$$\mathrm{HO^-} + \mathrm{CH_3CH_2I} \longrightarrow [\mathrm{HO} \cdots \underset{\mathrm{CH_3}}{\overset{\mathrm{H\ \ H}}{\mathrm{C}}} \cdots \mathrm{I}]^-$$

$$\longrightarrow \mathrm{HOCH_2CH_3} + \mathrm{I^-}$$

Nucleophiles other than the hydroxide ion that react in this fashion are ammonia and trimethylamine, and the ethoxide, acetate, cyanide, azide, acid sulfide, and hydride ions.

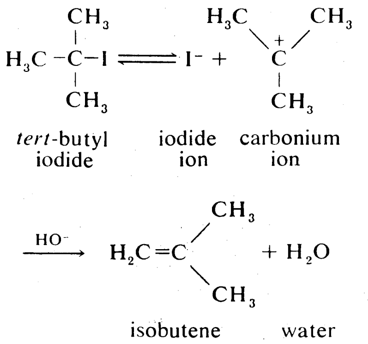
Elimination reactions

Whenever alkyl halides are hydrolyzed, a certain amount of olefin is formed by an elimination process. Elimination sometimes becomes the main reaction, as in the hydrolysis of tertiary halides, in which the carbonium ion formed can

Table 27: Melting Points and Boiling Points of Aryl Halides (°C)

compound	fluoro compound		chloro compound		bromo compound		iodo compound	
	melting point	boiling point	melting point	boiling point	melting point	boiling point	melting point	boiling point
Monohalobenzene	−41.9	85	−45.2	132.0	−30.6	156.2	31	188.5
o-Dihalobenzene	−34	91	−17.2	179.2	+ 6.7	225	26.7	286
m-Dihalobenzene	−59.3	83	−26.3	172	− 7	220	35	285
p-Dihalobenzene	−13	89	53.0	174.5	+87.3	220	129	285
1,2,3-Trihalobenzene	−13.5	95	52.4	218	88		116	
1,2,4-Trihalobenzene	−35	88	16.6	213	44		91.5	
1,3,5-Trihalobenzene	− 5.5	75.5	63	203	119	271	184	
1,2,3,4-Tetrahalobenzene	−42	95	47.5	254	47.5		136	
1,2,3,5-Tetrahalobenzene	−48	83	51	246	98		148	
1,2,4,5-Tetrahalobenzene	4	90	138	244	173		254	
Pentahalobenzene	−48	85	87	276	159		172	
Hexahalobenzene	5	80	229	326	326		340–350	

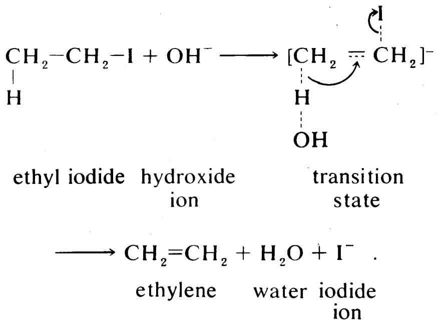
easily eliminate a proton. An example is the hydrolysis of *tert*-butyl iodide to give isobutene:



$$(CH_3)_3C{-}I \rightleftharpoons I^- + (CH_3)_3C^+ \xrightarrow{HO^-} H_2C{=}C(CH_3)_2 + H_2O$$

This is a unimolecular elimination—E_1 reaction—so classified because the rate at which it occurs is governed by the rate of formation of the carbonium ion from the alkyl halide.

Bimolecular eliminations—E_2 reactions—occur mainly with primary halides. They occur by removal of a hydrogen ion from the carbon atom adjacent to that carrying the halogen atom. For example, the elimination reaction involving ethyl iodide proceeds as shown (with the movements of the electrons comprising the various bonds indicated by curved arrows):



$$H{-}CH_2{-}CH_2{-}I + OH^- \longrightarrow [HO\cdots H\cdots CH_2 \cdots CH_2 \cdots I]^- \longrightarrow CH_2{=}CH_2 + H_2O + I^-\ .$$

By suitable choice of the reaction conditions (that is, by selecting the right solvent, temperature, and concentration of nucleophile) it is often possible to influence whether a particular reaction proceeds by bimolecular or unimolecular substitution or by an elimination reaction. Because the products of these reactions often are different, the distinction may be vitally important in a synthetic process.

Alkyl halides usually react with metals by a process that generates free radicals—that is, substances with single (unpaired) electrons. The free radical may then react with a second metal atom, resulting, as shown in the example below, in formation of an organometallic compound (that is, an organic compound with a metal substituent):

$$CH_3(CH_2)_3Br + Li\cdot \longrightarrow LiBr + CH_3(CH_2)_2\dot{C}H_2$$

n-butyl bromide   lithium metal   lithium bromide   *n*-butyl radical

$$\xrightarrow{Li\cdot} CH_3(CH_2)_3Li\ .$$

n-butyllithium

(In the names of these compounds *n*, for normal, indicates that the hydrocarbon chain is straight, or unbranched.) With magnesium, the reaction is carried out in an ether as solvent to give an organomagnesium halide:

$$R{-}I + Mg \longrightarrow R\cdot + \cdot MgI$$

alkyl iodide   magnesium   alkyl radical   iodo-magnesium radical

$$\longrightarrow R{-}Mg{-}I.$$

alkyl-magnesium iodide

In these reactions iodides and bromides are generally used because chlorides and fluorides react much less readily. The reduction of alkyl halides by sodium and alcohol, zinc and dilute acid, or similar reducing agents may also be a free-radical process:

$$RI + H\cdot \longrightarrow R\cdot + HI;$$

alkyl iodide   hydrogen atom   alkyl radical   hydrogen iodide

$$R\cdot + H\cdot \longrightarrow RH.$$

alkyl radical   hydrogen atom   alkane

Polyhalides. *Preparation.* Alkylidene halides, in which two halogen atoms are attached to the same carbon atom, can be made from aldehydes and ketones. For example, isopropylidene chloride is produced from acetone and phosphorus pentachloride; the reaction proceeds according to the following equation:

$$PCl_5 + CH_3COCH_3$$

phosphorus pentachloride   acetone

$$\longrightarrow POCl_3 + CH_3CCl_2CH_3.$$

phosphorus oxychloride   isopropylidene chloride

To prepare the fluorides, sulfur tetrafluoride or phenyl sulfur trifluoride is used with an aldehyde or ketone. Another general method for preparing alkylidene halides is to add a halogen acid to a vinyl halide or an alkyne. An example is the addition of hydrogen chloride to acetylene, vinyl chloride being produced first:

$$HC{\equiv}CH \xrightarrow{HCl} CH_2{=}CHCl \xrightarrow{HCl} CH_3CHCl_2\ .$$

acetylene   vinyl chloride   ethylidene chloride

Alkylene dihalides, in which two halogen atoms are attached to different carbon atoms, are formed from halogens and alkenes by electrophilic addition, or from glycols (compounds with two alcohol groups) by nucleophilic substitution exactly as the alkyl halides are prepared from ordinary alcohols.

Trihaloalkanes

The best known trihaloalkanes are those in which all three halogen atoms are on the same carbon atom; these are the haloforms, of which chloroform, $CHCl_3$, is the most familiar. Haloforms generally are made by the halogenation and subsequent alkaline hydrolysis of compounds containing the acetyl group (CH_3CO-) or structures capable of producing it on oxidation. Fluorination of acetophenone, for instance, followed by hydrolysis of the product, gives fluoroform, as shown:

$$C_6H_5COCH_3 \longrightarrow C_6H_5COCF_3$$

acetophenone   α,α,α-trifluoro-acetophenone

$$\xrightarrow{H_2O} C_6H_5COOH + HCF_3\ .$$

benzoic acid   fluoroform

Larger molecules can be made by the free-radical addition of chloroform to olefins, as follows:

$$CH_2{=}CH_2 + \cdot CCl_3 \longrightarrow \cdot CH_2CH_2CCl_3$$

ethylene   trichloromethyl radical (from $CHCl_3$)   intermediate radical

$$\cdot CH_2CH_2CCl_3 + HCCl_3$$

intermediate radical   chloroform

$$\longrightarrow CH_3CH_2CCl_3 + \cdot CCl_3.$$

1,1,1-trichloro-propane   trichloromethyl radical, regenerated

If the proportion of chloroform present in the reaction mixture is reduced, polymerization (formation of long, chainlike molecules with repeating units) occurs because the intermediate radical reacts with more olefin, giving a series of products of varying chain length, all of which have trichloromethyl end groups. Such products are formed as in the following equations:

$$\underset{\substack{\text{intermediate}\\ \text{radical}}}{\cdot CH_2CH_2CCl_3} + \underset{\text{ethylene}}{nCH_2{=}CH_2} \longrightarrow \underset{\text{polymer radical}}{\cdot CH_2CH_2(CH_2CH_2)_nCCl_3}$$

$$\underset{\text{polymer radical}}{\cdot CH_2CH_2(CH_2CH_2)_nCCl_3} + \underset{\text{chloroform}}{HCCl_3}$$

$$\longrightarrow \underset{\text{polymer product}}{CH_3CH_2(CH_2CH_2)_nCCl_3} + \underset{\substack{\text{trichloromethyl}\\ \text{radical}}}{\cdot CCl_3},$$

in which n is a variable number that depends on the number of times the first step is repeated.

Compounds more highly halogenated than the above classes of polyhalides result from the direct halogenation of alkanes by elemental fluorine, chlorine, or bromine. This process, too, is considered to involve free radicals. Because the energy liberated in replacing hydrogen atoms by fluorine atoms is more than enough to break a carbon-to-carbon bond, the vapour-phase fluorination of alkanes occurs violently unless the reaction is moderated by dilution of the reactants with gaseous nitrogen and unless the heat produced by the reaction is conducted away by a metal surface. As an alternative, metal fluorides can be used instead of elemental fluorine, or an electrochemical process can be employed. In most instances, but not in the electrochemical method, the amount of partly fluorinated material can be adjusted by varying the reaction conditions, but mixtures are always obtained. Chlorination of alkanes with elemental chlorine in the liquid phase gives mainly polychloro compounds but, in the gas phase, it is possible to arrange conditions so as to give more lightly chlorinated material. Bromination is carried out like chlorination; direct iodination, however, is not feasible. Fluorination is such a vigorous process that there is no discrimination in the attack on the different types of hydrogen atoms; in chlorination and, to a greater extent in bromination, tertiary hydrogen atoms are substituted more readily, followed by secondary and primary hydrogens, in order.

Direct halogenation

Reactions. The reactions of alkylene dihalides are very similar to the reactions of alkyl halides, but, when there are two or more halogen atoms on the same carbon atom, there is a sharp decrease in the reactivity of the halogen atoms. In general, as the atomic weight of the halogen atoms increases the stability of the polyhaloalkanes decreases, not only because the carbon–halogen bonds are weaker, but also because of the crowding caused by the packing together of the larger atoms. Stabilities decrease enormously, therefore, in the series from carbon tetrafluoride, through the chloride and bromide to carbon tetraiodide. Furthermore, 1,1,1-trifluoroethane is resistant to hydrolysis by acids, whereas the trichloro compound is easily hydrolyzed to acetic acid. Because the fluorine atom is not much bigger than the hydrogen atom, it is possible to prepare series of fluorocarbon derivatives entirely analogous to the related hydrocarbons. The sizes of the other halogen atoms, however, prevent the formation of many fully halogenated compounds. Fully fluorinated alkanes (called perfluoroalkanes) are particularly stable—both to heat and to chemical reagents—because of the very strong carbon–fluorine bonds and because the fluorine atoms are of just such size that they give good protection to the carbon chains.

Influences on reactivity

Polyhalogenated compounds mainly undergo elimination reactions. Halogen atoms are eliminated either as molecules of the halogen acid (HX) or as molecules of the free halogen itself (X_2). In both cases iodides are most reactive, and fluorides least. Elimination reactions are widely used in the preparation of halogenated alkenes and alkynes (see below *Alkenyl halides* and *Alkynyl halides*). Exchange of one kind of halogen atom for another occurs when polyhalo-compounds are heated with aluminum halides.

ALKENYL HALIDES

Preparation. Alkenyl halides, halogen derivatives of olefins, differ from one another in their chemical behaviour chiefly because of differences in the relationship of the halogen atom to the carbon–carbon double bond. Vinyl halides are those alkenyl halides in which a halogen atom is attached directly to one of the double-bond carbons; allyl halides are those in which the halogen is attached to one of the carbons adjacent to the double-bond carbons. Only in these classes of alkenyl halides is there interaction between the electrons of the halogen atoms and those of the double bond. Therefore, only these alkenyl halides show unique chemical properties; others of the general class resemble ordinary alkyl halides in their properties and methods of preparation.

Varieties of alkenyl halides

One method that is used to prepare vinyl halides of many different types is the addition of an appropriate simple molecule to an alkyne (triply bonded hydrocarbon). These reactions may occur spontaneously, when a simple molecule like a halogen acid is added, or they may need heating, when more complex molecules are to be added. Examples of both are shown below:

$$\underset{\text{acetylene}}{CH{\equiv}CH} + \underset{\substack{\text{hydrogen}\\ \text{chloride}}}{HCl} \longrightarrow \underset{\substack{\text{vinyl}\\ \text{chloride}}}{CH_2{=}CHCl};$$

$$\underset{\substack{\text{trifluoromethyl-}\\ \text{acetylene}}}{CF_3C{\equiv}CH} + \underset{\substack{\text{trichloro-}\\ \text{iodomethane}}}{CCl_3I}$$

$$\xrightarrow{\text{heat}} \underset{\substack{\text{1,1,1-trifluoro-2-iodo-}\\ \text{4,4,4-trichloro-2-butylene}}}{CF_3CI{=}CHCCl_3}$$

The other common methods of preparing vinyl halides are dehydrohalogenation (removal of hydrogen halides) and dehalogenation (removal of halogens) of polyhaloalkanes. Dehydrohalogenation can, in simple cases, be performed by heating with bases, for example:

$$\underset{\substack{\text{ethylene}\\ \text{dichloride}}}{CH_2ClCH_2Cl} + \underset{\substack{\text{sodium}\\ \text{hydroxide}}}{NaOH}$$

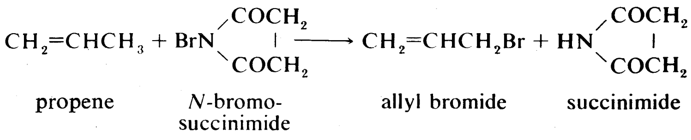
$$\longrightarrow \underset{\substack{\text{vinyl}\\ \text{chloride}}}{CH_2{=}CHCl} + \underset{\substack{\text{sodium}\\ \text{chloride}}}{NaCl} + \underset{\text{water}}{H_2O}.$$

The most common method of dehalogenation is to heat the haloalkane in ethanol solution with zinc dust, as in the following example:

$$\underset{\substack{\text{1-fluoro-1,2-}\\ \text{dibromoethane}}}{CH_2BrCHBrF} \xrightarrow[C_2H_5OH]{Zn} \underset{\text{vinyl fluoride}}{CH_2{=}CHF}.$$

Allyl halides are prepared by the action of the appropriate halogen acid on allyl alcohols. Allyl bromide also is made by treating propene with *N*-bromosuccinimide. The bromide can be converted into other halides by halogen exchange.

Allyl halides



$$\underset{\text{propene}}{CH_2{=}CHCH_3} + \underset{\substack{N\text{-bromo-}\\ \text{succinimide}}}{BrN\begin{matrix}\diagup COCH_2\\ \vert\\ \diagdown COCH_2\end{matrix}} \longrightarrow \underset{\text{allyl bromide}}{CH_2{=}CHCH_2Br} + \underset{\text{succinimide}}{HN\begin{matrix}\diagup COCH_2\\ \vert\\ \diagdown COCH_2\end{matrix}}$$

Reactions. In the vinyl halides, because of the interaction of the electrons of the halogen atom with those of the double bond, the carbon to halogen bond is stronger than it is in ordinary alkyl halides. Moreover, the carbon atom to which the halogen is attached does not carry a positive charge, as does the corresponding carbon atom in typical alkyl halides, and as a result nucleophilic replacement of the halogen atom in vinyl halides does not occur. This lack of positive charge, however, does not prevent the formation of organomagnesium halides, nor does it interfere with normal electrophilic and free radical additions to the double bond.

Polyhalo vinyl compounds

As the number of halogen atoms in a vinyl halide molecule increases, the tendency to undergo electrophilic additions decreases and that to undergo nucleophilic additions increases. Both effects are due to the decrease in electron density of the double bond caused by the electronegative substituents. For polyhalo vinyl compounds, then, the most commonly observed reaction is the addition of nucleophiles, such as alcohols, amines, and thiols. For example, the addition of ethanol to 1,1,2-trichloroethylene gives an ether as an initial reaction product; this is subsequently dehydrohalogenated to dichlorovinyl ethyl ether.

$$\underset{\text{1,1,2-trichloroethylene}}{CCl_2{=}CHCl} + \underset{\text{ethanol}}{C_2H_5OH} \longrightarrow \underset{\text{initial product}}{C_2H_5OCCl_2CH_2Cl}$$

$$\xrightarrow{NaOC_2H_5} \underset{\text{dichlorovinyl ethyl ether}}{C_2H_5OCCl{=}CHCl}$$

Allyl halides, which conform to the structural pattern $CH_2{=}CHCH_2X$, are very much more susceptible to nucleophilic substitution of the halogen atom than are alkyl halides. When a carbonium ion intermediate is formed, (S_N1 reaction) it may be represented by the two formulas:

$$\underset{\gamma}{CH_2}{=}CH{-}\underset{\alpha}{\overset{+}{CH_2}} \longleftrightarrow \underset{\gamma}{\overset{+}{CH_2}}{-}CH{=}\underset{\alpha}{CH_2}.$$

Attack by the nucleophile occurs at either the alpha (α) or gamma (γ) carbon atoms (as shown). This choice becomes significant for unsymmetrical allyl carbonium ions as the product is different in each case. Bimolecular (S_N2) substitution usually involves direct replacement of the allyl halogen atom, but this reaction also can occur with attack at the gamma position to give an alternate product.

Halogenated alkenes of only two carbon atoms usually polymerize readily, but halogenated alkenes bearing large, halogenated alkyl groups attached to the unsaturated carbon atoms may not do so. Fluorinated olefins easily form cyclic dimers when heated above 200° C (392° F). Halogenated alkenes often can be dehydrohalogenated or dehalogenated to alkynes.

ALKYNYL HALIDES

Preparation. In many cases, compounds containing halogen atoms directly joined to triply bonded carbon atoms are formed by reaction between molecular halogen and a metallic derivative of the alkyne. Various sodium acetylides, for example, react with free iodine to give the corresponding iodo compounds.

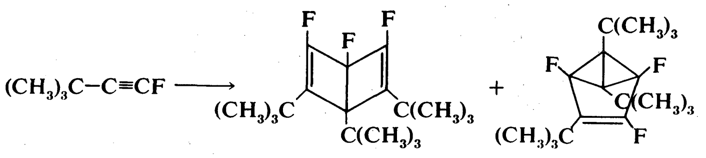
$$RC{\equiv}CNa + I_2 \longrightarrow RC{\equiv}CI + NaI$$

In other instances, alkynyl halides are more easily prepared by dehalogenation or dehydrohalogenation reactions of halogenated alkanes of alkenes. Examples of both reactions, using1,1,2-trichloroethylene as starting material, are shown below:

$$\underset{\text{chloroacetylene}}{CCl{\equiv}CH} \xleftarrow[\text{alcohol}]{\text{Zn}} \underset{\text{1,1,2-trichloroethylene}}{CCl_2{=}CHCl} \xrightarrow[\text{NaOH}]{\text{aq alc}} \underset{\text{dichloroacetylene}}{CCl{\equiv}CCl.}$$

Instability of halo-alkynes

Reactions. Halogenated acetylenes (alkynes), in which the halogen atom is directly bonded to one of the unsaturated carbon atoms, generally are unstable; the lower the molecular weight, the more likely they are to explode. This instability may be due to spontaneous exothermic (heat-producing) polymerization, which, in some cases, can be regulated so that trimerization to halogenated benzenes occurs. Sometimes isomers of benzene derivatives are formed, as for example, in trimerization of *tert*-butylfluoroacetylene:



tert-butylfluoroacetylene — isomeric benzene derivatives

In these structures the polygons represent rings of carbon atoms (the carbon atoms at the angles not being shown), with the lines indicating the bonds between atoms.

A halogen atom attached to an unsaturated carbon atom in a halogenated acetylene is resistant to nucleophilic substitution. Most nucleophiles, therefore, react with haloalkynes first by addition and then by elimination of hydrogen halide. Ammonia, for instance, adds to dichloroacetylene; the initial product then loses hydrogen chloride to give an alkyne, and this in turn isomerizes to a nitrile.

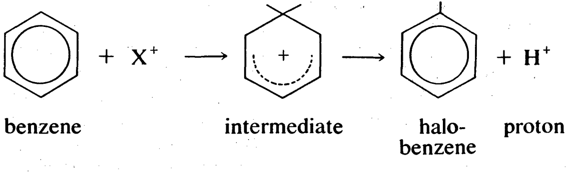
$$\underset{\text{dichloroacetylene}}{ClC{\equiv}CCl} + \underset{\text{ammonia}}{NH_3} \longrightarrow \underset{\text{1,2-dichloroaminoethylene}}{ClCH{=}C(NH_2)Cl}$$

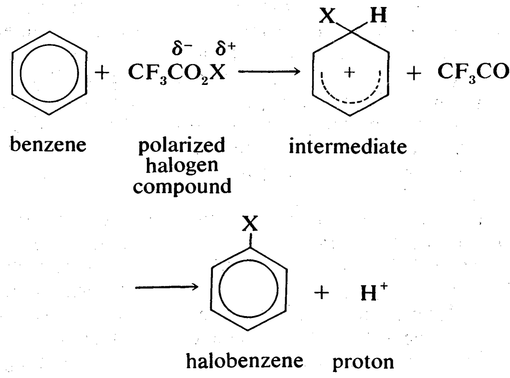
$$\longrightarrow \underset{\text{hydrogen chloride}}{HCl} + \underset{\text{1-amino 2-chloroacetylene}}{ClC{\equiv}CNH_2} \xrightarrow{\text{isomerization}} \underset{\text{chloroacetonitrile}}{CH_2ClC{\equiv}N}$$

In most solvents, such as carbon tetrachloride, electrophilic additions of halogens and halogen acids to alkynyl halides take place, but the reaction is vigorous and must be controlled carefully. The hydrogen in monohaloacetylenes is acidic, as in acetylene itself, and can be replaced by metals. Haloacetylenes in which the halogen is remote from the triple bond are generally stable and show the separate reactivities of alkynes and alkyl halides. The perfluoroalkylalkynes are less unstable than most other alkynes, and they retain the reactivity of the triple bond, readily undergoing free-radical, electrophilic, and nucleophilic addition reactions.

ARYL HALIDES

Nuclear-halogenated aromatic compounds. *Preparation.* There are three main methods for producing aromatic compounds that carry halogen atoms on the aromatic ring (or aromatic nucleus, as it is often called). These are (1) replacement of hydrogen by halogen, (2) replacement of an amino group by halogen, and (3) dehydrohalogenation or dehalogenation of polyhalocyclohexanes or polyhalocyclohexenes.





Replacement of hydrogen

The mechanism of hydrogen replacement is one of electrophilic substitution. It involves, first, the production of either a positively charged halogen ion or a polarized molecule in which the halogen has a partial positive charge (the electrophile). This positive centre then reacts with benzene to form an intermediate from which a proton is subsequently eliminated. Ferric chloride, aluminum chloride, and boron trifluoride are commonly used as catalysts with chlorine or bromine to produce the positively charged halogen ion.

$$\mathrm{Cl_2 + AlCl_3 \rightleftharpoons Cl^+ + AlCl_4^-}$$

Iodination requires the presence of an oxidizing agent, such as nitric acid, to remove hydrogen iodide from the reaction as soon as it is produced, as well as to encourage the formation of positively charged iodine. Hypohalous acids, *N*-chloro- or *N*-bromoamides, and acyl hypohalites are useful halogenating agents because they are polarized in such a way that the halogen atom carries a positive charge.

A chlorine, bromine, or iodine atom in the benzene nucleus deactivates it and causes further substitution to occur in the *ortho* and *para* positions (see above *General chemical properties*). Fluorine also causes *ortho* and *para* orientation, but does not deactivate the nucleus. Excellent yields of nuclear halogenated compounds are formed by chlorinating, brominating, or iodinating substituted benzenes.

Replacement of amino groups

Any of the four halogens can be introduced into the nucleus by diazotizing an aromatic amine and then allowing the resulting diazonium salt to decompose in the presence of a suitable salt of the halogen. Chlorine and bromine, for example, are introduced by heating a diazonium salt in dilute acid solution with cuprous chloride or bromide:

$$Ar\mathrm{NH_2} \longrightarrow Ar\overset{+}{\mathrm{N}}{\equiv}\mathrm{N\ X^-} \xrightarrow[\mathrm{HX}]{\mathrm{Cu_2X_2}} Ar\mathrm{X} + \mathrm{N_2}.$$

aromatic amine	diazonium salt	aryl halide	nitrogen

Ar = any aromatic nucleus

Alternatively, the appropriate diazonium salt can be heated in dilute acid with copper powder:

$$Ar\overset{+}{\mathrm{N}}{\equiv}\mathrm{N\ X^-} \xrightarrow[\mathrm{HX}]{\mathrm{Cu}} Ar\mathrm{X} + \mathrm{N_2}.$$

diazonium salt	aryl halide	nitrogen

Iodine is introduced simply by warming a diazonium salt in the presence of aqueous potassium iodide, for example:

$$\mathrm{C_6H_5\overset{+}{N}{\equiv}N\ Cl^- + KI \longrightarrow C_6H_5I + N_2 + KCl.}$$

benzenediazonium chloride	potassium iodide	iodobenzene	nitrogen	potassium chloride

The most common method used to introduce fluorine is the dry distillation of a diazonium fluoroborate:

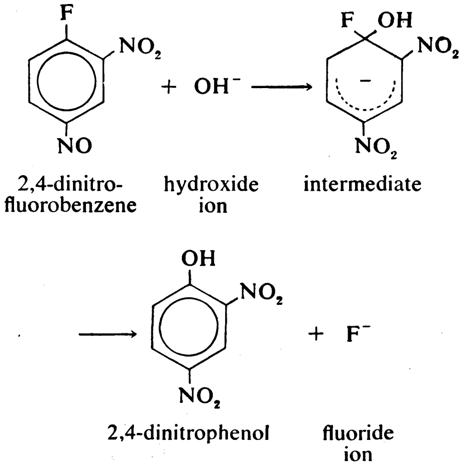
$$Ar\overset{+}{\mathrm{N}}{\equiv}\mathrm{NBF_4^-} \longrightarrow Ar\mathrm{F} + \mathrm{N_2} + \mathrm{BF_3}.$$

diazonium fluoroborate	aryl fluoride	nitrogen	boron trifluoride

The photochemical (*i.e.*, free-radical) chlorination of benzene causes the addition of chlorine to the double bonds and gives a mixture of hexachlorocyclohexanes. Dehydrohalogenation of this mixture by heating with aqueous alcoholic alkali gives a mixture of trichlorobenzenes.

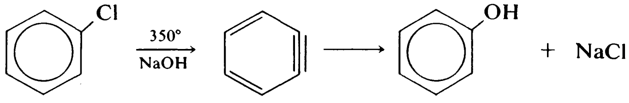
From the mixture obtained by fluorinating benzene with cobalt trifluoride, octa- and nonafluorocyclohexanes can be isolated. Dehydrofluorination of these substances by heating with concentrated aqueous potash gives penta- and hexafluorobenzenes. Deca-, nona-, and octafluorocyclohexanes can be dehydrofluorinated to polyfluorocyclohexenes and -dienes, and these compounds in turn can be defluorinated giving hexa-, penta-, and tetrafluorobenzenes.

Reactions. The main difference between aryl halides and alkyl halides is that the former are much more resistant to nucleophilic attack. If, however, the halogen in an aromatic compound is activated by the presence of other groups in the molecule, or if especially vigorous conditions (such as elevated temperatures and the presence of strong bases) are used, then nucleophilic substitutions or elimination–addition reactions (see below), which give the same end result, can take place. Thus, 2,4-dinitrofluorobenzene on treatment with a nucleophile such as hydroxide ion undergoes a bimolecular substitution to give a phenol. When the nucleus is not activated either stronger bases or much more vigorous conditions are needed for nucleophilic attack; in such cases elimination of hydrogen halide gives a benzyne structure, to which the nucleophile then adds.



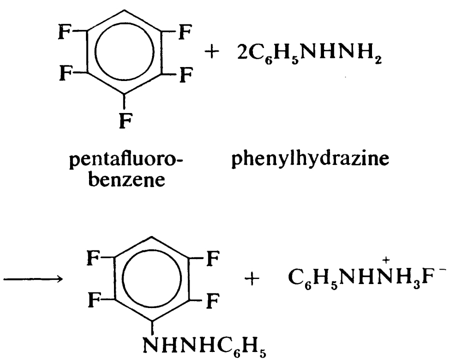
2,4-dinitrofluorobenzene — hydroxide ion — intermediate

2,4-dinitrophenol — fluoride ion



chlorobenzene — benzyne — phenol — sodium chloride

In hexachlorobenzene, the halogen atoms activate the nucleus in the same way as do the nitro groups in dinitrofluorobenzene; and, on heating with potassium fluoride, hexafluoro- and mixed chlorofluorobenzenes are formed by nucleophilic reactions. Pentafluoro- and chlorofluoropyridines are similarly made. Polychloro- and polyfluorobenzenes are highly susceptible to this form of attack. Reagents such as ammonia, sodium methoxide, lithium aluminum hydride, and phenylhydrazine often are used as nucleophiles. An example is given below:



pentafluorobenzene — phenylhydrazine

2,3,5,6-tetrafluorohydrazobenzene — phenylhydrazine hydrofluoride

The nuclear hydrogen atoms in aryl halides generally are subject to electrophilic attack and these compounds, therefore, undergo such typical aromatic reactions as halogenation, nitration, and sulfonation. Substitution occurs chiefly in the *para* position, but also in the *ortho* position.

Reactions of aryl halides with metals are similar to those of alkyl halides. Aryl fluorides do not form organomagne-

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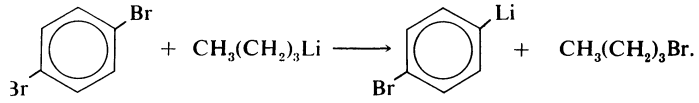
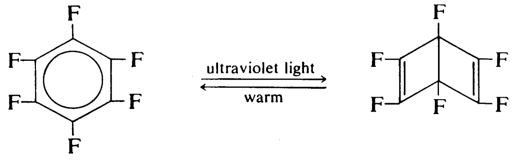
sium halides, and chlorides react sluggishly; bromides and iodides, however, react readily. In polyhalo compounds, any hydrogen atoms left in the molecule are acidic because of the presence of the electronegative halogen atoms, and as a result these compounds regularly form organometallic compounds by hydrogen–metal exchange. Organolithium compounds often are formed by halogen–metal exchange in the less halogenated compounds. For example:



p-dibromobenzene butyllithium *p*-bromophenyllithium butyl bromide

In nuclear halogenated compounds, chlorine, bromine, and iodine atoms are replaced by hydrogen atoms when the compounds are heated with hydrogen gas in the presence of a catalyst such as finely divided nickel, platinum, or palladium. Hexafluorobenzene, on ultraviolet irradiation, is converted into the so-called Dewar isomer (an abnormal form of benzene), which reverts to the normal form on gentle heating, but which may explode.



hexafluorobenzene Dewar isomer

ide-chain
aloge-
ation

Side-chain-halogenated aromatic compounds. *Preparation.* Direct halogenation of an alkyl group (side chain) attached to an aromatic ring is a free-radical process and, as such, it is helped by exposure to visible or ultraviolet light and by boiling the reaction mixture, which aids dissociation of the halogen molecules to free atoms. After substitution is complete, the reagents effecting halogenation also cause saturation of the nuclear double bonds. In side-chain halogenation it is essential to avoid any trace of catalysts that promote nuclear substitution. Fluorine reacts so vigorously that it cannot be introduced preferentially into the side chain in this way, and iodine requires the presence of an oxidizing agent. The method is ideal for chlorine and bromine.

$$Br_2 \xrightarrow[\text{CS}_2 \text{ or CCl}_4 \text{ soln.}]{\text{ultraviolet light}} 2Br\cdot;$$

molecular bromine atomic bromine

$$C_6H_5CH_3 + Br\cdot \xrightarrow[\text{or CCl}_4]{\text{CS}_2} C_6H_5CH_2\cdot + HBr;$$

toluene bromine atom benzyl radical hydrogen bromide

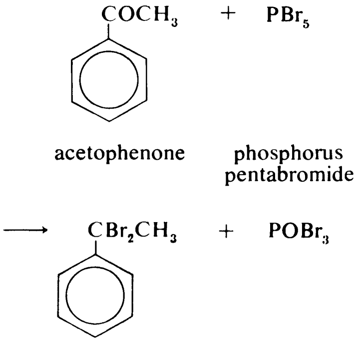
$$C_6H_5CH_2\cdot + Br_2 \longrightarrow C_6H_5CH_2Br + Br\cdot$$

and so on.

Sulfuryl chloride, with a trace of benzoyl peroxide as a free-radical initiator, can be used as a source of chlorine atoms for side-chain chlorination. Such chlorinations proceed rapidly in the dark. Other reagents that produce halogen atoms and therefore start chain reactions are tertiary butyl hypochlorite and hypoiodite and *N*-bromosuccinimide. The degree of halogenation can be controlled by regulating the amount of reagent used, but mixtures are always produced.

Other methods for preparing compounds halogenated in the side chain are similar to those used to prepare alkyl halides, either from the corresponding alcohol, such as benzyl alcohol ($C_6H_5CH_2OH$), or by the electrophilic addition of hydrogen halides or halogens to aryl olefins.

Aromatic aldehydes and ketones react with phosphorus halides or with sulfur tetrafluoride to give dihalo compounds.



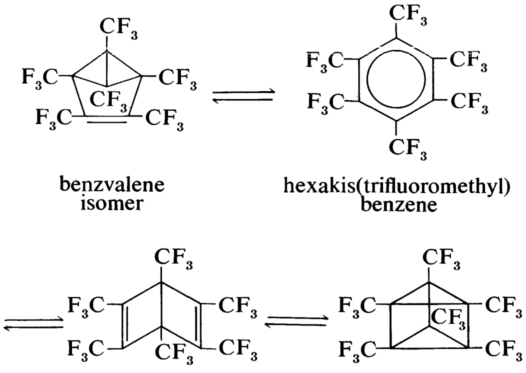
acetophenone phosphorus pentabromide

α,α-dibromoethylbenzene phosphorus oxybromide

Aromatic acids react with sulfur tetrafluoride to give trifluoromethyl derivatives.

Reactions. Typical examples of aromatic compounds with halogen atoms in their side chains are benzyl chloride ($C_6H_5CH_2Cl$), benzylidene chloride ($C_6H_5CHCl_2$), and benzotrichloride ($C_6H_5CCl_3$). The reactivities of these compounds are intermediate between those of the alkyl and allyl halides with similar halogen atoms. Thus, benzyl halides react readily with nucleophiles; hydrolysis of benzylidene halides gives benzaldehyde, and that of benzotrichloride gives benzoic acid. When halogen atoms are not attached to the carbon atom adjacent to the benzene nucleus, the compounds react almost like similar alkyl halides.

Hexakis (trifluoromethyl) benzene undergoes ready photochemical valence bond isomerization to the benzvalene, "Dewar," and prismane isomers, which are surprisingly stable for compounds of this type.

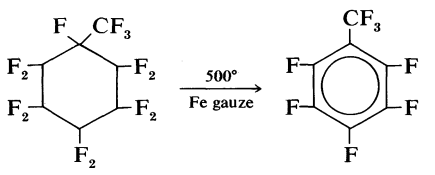


benzvalene isomer hexakis(trifluoromethyl) benzene

Dewar isomer prismane isomer

Nuclear- and side-chain-halogenated aromatic compounds. *Preparations.* Compounds that are lightly halogenated in both nucleus and side chain can obviously be prepared by methods already described. Of those that are highly or completely halogenated, only the polychloro and polyfluoro compounds are known, presumably because bromine and iodine atoms are too large. The photochemical chlorination of the side chains of benzene homologues fully chlorinated in the nucleus proceeds, for example, to the dichloromethyl or the trichloroethyl stage, where it stops. It also is difficult to complete the nuclear chlorination of benzene homologues with fully chlorinated side chains. There is one catalyst, aluminum trichloride and sulfur monochloride with sulfuryl chloride, sufficiently active to give perchloroalkaryl compounds (the prefix per- indicating complete substitution) from starting materials with fully halogenated side chains and nonhalogenated nuclei. Of the simple benzene homologues, only *m*-xylene and mesitylene resist complete chlorination by this method. Perfluoro homologues of benzene are prepared by first converting the hydrocarbons (using cobalt trifluoride) to perfluorocyclohexanes. The latter, then, are defluorinated to the fully fluorinated homologue of benzene.

Fully halogenated compounds



perfluoromethyl cyclohexane
octafluorotoluene

Nuclear fluorinated aromatic compounds can be fluoroalkylated by nucleophilic attack with anions generated from fluoride ion and perfluoroalkenes, as shown below:

$$CF_3CF{=}CF_2 + F^- \longrightarrow CF_3\bar{C}FCF_3;$$

perfluoropropene · fluoride ion · intermediate anion

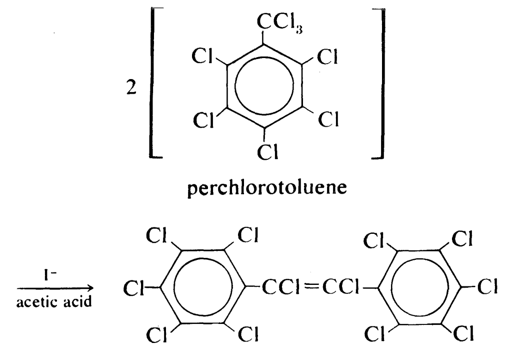
$$C_6F_6 + CF_3\bar{C}FCF_3$$

hexafluorobenzene · anion

$$\longrightarrow C_6F_5CF(CF_3)_2 + F^-.$$

perfluorocumene · fluoride ion

The behaviour of alkylbenzenes lightly halogenated in the nucleus and side chain is predictable on the basis of the separate effects of the two classes of halogen atoms. With perhalogenated compounds, new properties appear. For example, prolonged treatment of perchloro compounds with chlorine, in the presence of a catalyst or ultraviolet light, causes chlorinolysis of the side chain to give hexachlorobenzene and the fully chlorinated alkane. Treatment of perchloro compounds with iodide ion in acetic acid causes dechlorination and sometimes coupling (union of two molecules); for example,

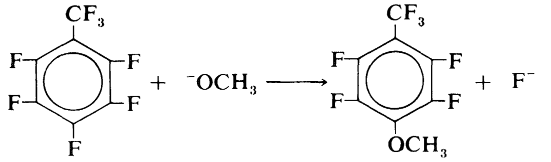


perchlorotoluene

perchlorostilbene

Hydrolysis of side-chain halogens

Chlorinated groups adjacent to the chlorinated benzene nucleus are easily hydrolyzed by heating with concentrated sulfuric acid: $C_6Cl_5CCl_2CCl_3 \rightarrow C_6Cl_5COCCl_3$. Trifluoromethyl groups in perfluoro- or chlorofluorobenzene homologues are similarly hydrolyzed to carboxyl groups. In all of these polyhalo compounds, but particularly in the fluorinated ones, the halogen atoms in the nucleus are susceptible to substitution by nucleophiles, such as lithium aluminum hydride, hydrazine, ammonia, methyllithium, and sodium alkoxides. In the perfluoro compounds, substitution occurs *para* to the perfluoroalkyl group (that is, in the position directly across the benzene ring).



perfluorotoluene · methoxide ion · *p*-methoxyperfluorotoluene · fluoride ion

Polyvalent iodine compounds (preparation and reactions). *Dichloroiodoarenes.* The iodine atom in an aryl iodide can exist in higher states of oxidation; *i.e.,* it can show valencies greater than one. Thus, when chlorine is bubbled into a solution of aryl iodide in cold dry chloroform, a dichloroiodoarene usually crystallizes out as a yellow solid. The formation of dichloroiodobenzene from iodobenzene occurs as follows:

$$C_6H_5I + Cl_2 \longrightarrow C_6H_5ICl_2.$$

These compounds, in which the iodine has a valence of three, are relatively unstable and decompose on heating to about 110° C (230° F). In the case of dichloroiodobenzene the product is *p*-chloroiodobenzene. Dichloro- and difluoroiodoarenes have been used as chlorinating and fluorinating agents.

Iodosyl compounds. When the dichloroiodoarenes are treated with aqueous alkali or aqueous pyridine, compounds containing the iodosyl group (—IO) are formed.

$$C_6H_5ICl_2 + 2NaOH \longrightarrow 2NaCl + H_2O + C_6H_5IO$$

dichloroiodobenzene · sodium hydroxide · sodium chloride · water · iodosylbenzene

Alternatively, these iodosyl compounds can be prepared directly from aryl iodides by oxidation. They behave as though they were anhydrides of the unknown bases, $ArI(OH)_2$, the salts of which can be made by dissolving the neutral iodosyl compounds in the appropriate acids (nitric, hydrochloric, hydrofluoric, and acetic).

Iodosyl compounds are oxidizing agents and liberate iodine quantitatively from acidified potassium iodide solution as follows:

$$C_6H_5IO + 2HI \longrightarrow C_6H_5I + I_2 + H_2O.$$

Iodyl compounds. Disproportionation (spontaneous oxidation and reduction) of iodosyl compounds on boiling with water or, more slowly, on standing, gives iodyl compounds and simple aryl iodides:

$$2C_6H_5IO \longrightarrow C_6H_5I + C_6H_5IO_2.$$

Iodyl compounds also can be made by oxidizing aryl iodides or iodosyl compounds.

Like iodosyl compounds, iodyl compounds are oxidizing agents that are quantitatively reduced by aqueous hydrogen iodide. They must be handled carefully because they explode on impact or on heating.

Iodonium compounds. A mixture of iodyl- and iodosylbenzene, on heating with silver oxide in aqueous suspension, gives a strongly alkaline solution of diphenyliodonium hydroxide:

$$2C_6H_5IO + 2C_6H_5IO_2 + Ag_2O + H_2O$$
$$\longrightarrow 2(C_6H_5)_2I^+\,OH^- + 2AgIO_3\ .$$

The addition of a soluble iodide to a solution of a diphenyliodonium hydroxide yields a precipitate of a diphenyliodonium iodide. Stable salts of iodonium hydroxides can be formed by neutralization with acids. Such salts react with a great variety of nucleophiles, to give nuclear substituted products and aryl iodides:

$$(Ar-\overset{+}{I}-Ar)X^- + OH^-$$
$$\longrightarrow ArOH + ArI + X^-\ .$$

They also decompose to substituted arenes on heating. For example:

$$(p\text{-}CH_3OC_6H_4-\overset{+}{I}-C_6H_5)\overset{-}{Br} \xrightarrow{180^\circ} p\text{-}CH_3OC_6H_4I + C_6H_5Br.$$

p-methoxyphenylphenyliodonium bromide · *p*-iodomethoxybenzene · bromobenzene

TECHNICAL AND ANALYTICAL ASPECTS OF ORGANIC HALOGEN COMPOUNDS

Use as solvents

Commercial uses. The polychloroalkanes and -alkenes are commonly used as solvents for drycleaning, chiefly because of their nonflammability. The most common are carbon tetrachloride and the trichloro- and tetrachloroethylenes. Chloroform is used as a solvent under controlled conditions, but its anesthetic properties make it dangerous. Because of their heavy vapours and their nonflammability some organic halogen compounds (carbon tetrachloride and dibromodifluoromethane, especially) have been used as fire extinguishers. Chloro compounds can be dangerous because under these conditions they give phosgene ($COCl_2$), a poisonous gas. The simple chlorofluoromethanes and -ethanes are used as refrigerants (Freons).

Chloroform was used for many years (mixed with acetone and ether) as an anesthetic. It is dangerous, however, because of the low margin between the lethal dose and the anesthetic dose and because the substance can damage the liver. It has now been completely superseded by fluothane ($CF_3CHBrCl$). Ethyl chloride is used as a local anesthetic.

Use as polymers

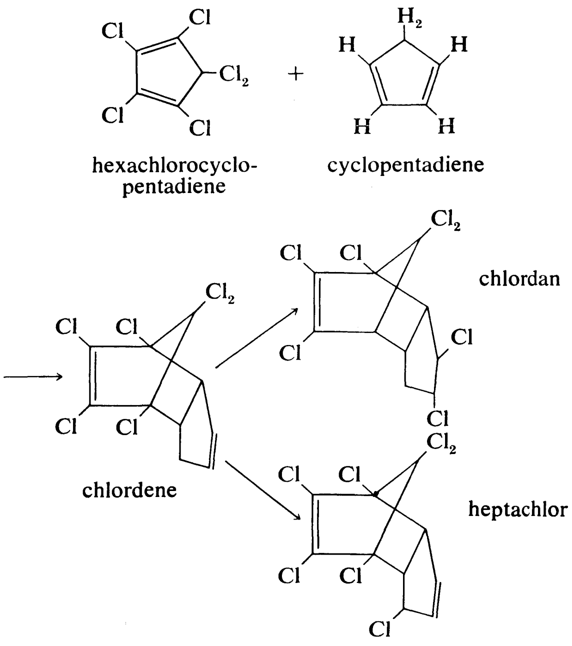
One of the most common polymers is polyvinyl chloride. It is used mainly for electrical insulation and artificial fabrics for rainwear and upholstery. Tetrafluoroethylene and chlorotrifluoroethylene give Teflon (Fluon) and Kel-F polymers, respectively; these have very high thermal stability and electrical resistance. Because of the low intermolecular forces between highly fluorinated compounds, the lower molecular weight Kel-F polymers are used as lubricants, and Teflon is used to pack bearings in motors. Copolymers of hexafluoropropene and vinylidene fluoride (Vitons) are among the most thermally stable elastomers. Electronic apparatus and switchgear are sometimes immersed in fully fluorinated liquid alkanes (Flutec PP5) and cycloalkanes (Flutec PP3), which conduct away the heat developed during operation and act as insulators.

Use as insecticides

Polychloro compounds are most important as insecticides. One of the earliest was DDT (dichlorodiphenyltrichloroethane), which can be made by condensing together chlorobenzene and chloral:

$$2C_6H_5Cl + CCl_3CHO$$
$$\xrightarrow{H_2SO_4} (p\text{-}ClC_6H_4)_2CHCCl_3 + H_2O.$$

Gammexane, the gamma isomer of hexachlorocyclohexane, has properties and uses similar to those of DDT. It is made by the free radical addition of three molecules of chlorine to benzene. Of the nine possible isomers, eight are formed, and—of these—only the gamma form is active. Another important group of insecticides is derived from hexachlorocyclopentadiene. Addition of cyclopentadiene to the hexachloro compound gives chlordene. This substance can be converted either to chlordane by addition of chlorine or to heptachlor by substitution of chlorine for a hydrogen atom:



Use of norbornadiene rather than cyclopentadiene in the above reaction gives aldrin, which can be converted to its epoxide dieldrin. Two similar insecticides, prepared by a slightly different route, are isodrin and endrin. The use of these highly chlorinated compounds has been criticized and curtailed because their stability causes them to accumulate in fish, birds, and animals. Monofluoroacetamide has been used as a systemic insecticide to kill black flies on plants such as roses; sodium monofluoroacetate is used as a rodenticide. Both are dangerous to human beings. Their volatile derivatives (*e.g.,* fluoroacetic acid and ethyl fluoroacetate) were among the first nerve gases.

Use as herbicides

Chlorinated and fluorinated benzenoid and heterocyclic compounds, and also urea derivatives, are used as herbicides. The reaction of chloral and urea gives dichloralurea, which is used as a soil sterilant to prevent the growth of weeds. The substance 3,5-dichloro-4-hydroxydifluoropyridine (Teklon) is used for the control of wild oats and couch grass.

The well-known herbicides 2,4-D and 2,4,5-T (respectively, 2,4-dichloro- and 2,4,5-trichlorophenoxyacetic acids) are used to kill broad-leaved plants and smaller shrubs, respectively.

The compound 5-fluorouracil is thought to replace uracil in nucleic acids and hence interfere with the biochemical operation of living cells. It is used in the treatment of some forms of cancer. The introduction of one or two fluorine atoms into steroid molecules either by replacing hydrogen or by addition across a double bond often enhances the hormonal activity of the steroid. Certain halogenated steroids also are used to reduce skin inflammation.

Separation, identification, and analysis. The introduction of a bromine or iodine atom into a hydrocarbon molecule increases the boiling point of the substance by about 100° to 150° C (180° to 270° F) and creates bonds that are less stable than carbon–hydrogen bonds. In the main, however, alkyl halides, aryl halides, and most polyhalo compounds are stable. The most common are gases or liquids, and the most usual way to purify them is by fractional distillation. Unfortunately, however, polyfluoro compounds have a pronounced tendency to form constant-boiling mixtures, and—as a result—they are often not separable in this way. The technique known as vapour phase chromatography has proved useful in this case, and in fact this method is now used extensively for the separation of mixtures of reasonably volatile halo compounds (with boiling points up to 200° C) of all kinds. Fundamentally, vapour phase chromatography is very similar to fractional distillation, the vapour of the mixture in a stream of nitrogen being adsorbed on and desorbed from a column of nonvolatile solvent supported on a substance like kieselguhr. By varying this stationary phase (dinonyl phthalate, silicone oil, and fluorolube are common materials used) it is usually possible to separate any mixture.

Spectroscopic identification

Apart from the use of melting and boiling points, pure organic halogen compounds are identified most often spectroscopically. There are strong characteristic absorptions in the infrared arising from stretching vibrations of the carbon-to-halogen bonds. Because the frequencies at which these absorptions occur are influenced by neighbouring groups, they give valuable information about the molecular structure, revealing, for example, whether the halide is an alkyl or aryl derivative. In the mass spectrometer, which separates materials on the basis of their relative masses, organic halides are usually sufficiently stable to show an ion derived from the intact molecule. As a result, an accurate molecular weight can easily be found. The breakdown pattern of the ions often gives strong clues to the structure of the parent compound.

Nuclear magnetic resonance spectroscopy, which measures the energy required to change the alignment of magnetic nuclei in a magnetic field, is used to determine the structures of compounds containing hydrogen and fluorine atoms. The resonance peaks associated with these atoms appear at different locations in the spectra of various compounds depending on the atoms adjacent to them. By finding the number of atoms in each different environment, the structure of a compound often can be determined.

By measuring another atomic property, the nuclear quadrupole moment, the structures of compounds containing chlorine and bromine can often be deduced. It is probable, however, that in the future the structures of chlorine-, bromine-, and iodine-containing compounds will best be determined by X-ray induced electron emission spectroscopy. This technique is based on the magnetic or electrostatic analysis of the electrons that are emitted from a substance on irradiation with X-rays. The energies of these electrons are characteristic of the atom from which they are emitted and of the environment of that atom. Hence, the results can be used both for elemental analysis and for structure determination.

Elemental analysis

Conventional elemental analysis of halogens in organic compounds involves, first, fusion of the compound with sodium or potassium, and then estimation of the alkali metal halides. Fluoride can be determined by weight as lead chlorofluoride, and chloride by titrating the alkali liberated by its action on mercuric oxycyanide. Bromide is quantitatively oxidized to bromate by sodium hypochlorite, and this, in turn, is used to liberate iodine from hydrogen iodide. In a similar fashion, iodide is converted to iodate by hypobromite, and this, again, is used to liberate iodine from hydrogen iodide. In both cases the free iodine is titrated with standard sodium thiosulfate.

(W.K.R.M.)

Organic nitrogen compounds

Since, by definition, organic compounds are compounds containing carbon, organic nitrogen compounds are substances the molecules of which contain at least one carbon and one nitrogen atom. Several hundred thousand such compounds are known, and new ones are being prepared or discovered continually. Organic nitrogen compounds are present in all known living organisms, as essential and often major components. Flesh, hair, horn, milk solids, and blood, for example, are composed largely of organic nitrogen compounds. Many drugs and medicinal agents, both natural and synthetic, including most narcotics, local anesthetics, the sulfa drugs, penicillin; most explosives, such as nitroglycerine and TNT (2,4,6-trinitrotoluene); some rocket fuels; many dyes; and some synthetic polymers (substances of high molecular weight made up of a number of identical or similar groups of atoms bonded together), such as nylons and the melamine resins, belong to this class.

No single characterization can be made to apply to the properties of organic nitrogen compounds. Some are gases—*e.g.,* hydrogen cyanide (formula HCN) and methylamine (CH_3NH_2); some liquids—*e.g.,* nitroglycerin

$$[O_2NOCH_2CH(ONO_2)CH_2ONO_2]$$

and aniline ($C_6H_5NH_2$); but most are solids. Some mix completely with water; others are quite insoluble. They may be acidic, basic, or neutral, coloured or colourless, volatile or nonvolatile, intensely poisonous or essential to life.

BONDING IN NITROGEN AND ITS COMPOUNDS

Bonding characteristics of nitrogen. The wide variation in properties of members of the group is directly related to the ability of nitrogen to take part with other elements in a large number and variety of molecular structures. This ability results from the capacity of the nitrogen atom to have a large number of so-called oxidation states.

Oxidation state

The electron pairs of covalent bonds are not shared equally when the bonded atoms are different, a fact that gives rise to the concept of oxidation state (or oxidation number), which is defined as the number of bonds shared with electronegative elements (*i.e.,* those to the right in the periodic table) minus the number of bonds shared with electropositive elements (*i.e.,* those to the left in the periodic table). The oxidation number of nitrogen ranges from −3, as in ammonia (NH_3), through zero, as in elemental nitrogen (N_2), to +5, as in nitric acid (HNO_3), and nitrogen occurs in known organic compounds in these and all intermediate oxidation states (−2, −1, +1, +2, +3, +4).

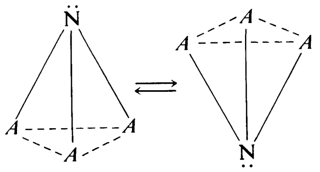
Organic nitrogen compounds are classified on the basis of the inorganic (*i.e.,* not containing carbon) constituent that can be considered the parent compound; *i.e.,* the compound from which the organic compound is produced by removing a hydrogen atom and replacing it with a group of atoms containing carbon. Nitrogen occurs in nature bound to hydrogen, oxygen, carbon, and other nitrogen atoms; it can also form bonds with other elements, including sulfur, phosphorus, fluorine, chlorine, bromine, and iodine.

Nitrogen-compound structures. In almost all organic nitrogen compounds, the three additional electrons required to fill the octet are obtained by sharing an electron pair with another atom, which contributes one electron to the shared pair, thus forming a covalent bond. Three such bonds, adding three electrons to the normal five, result in a filled octet; the fourth pair of electrons of the nitrogen atom is not shared. An example is the reaction shown below (in which *A* represents any atom with one electron available to share, and each electron is represented by a dot); one nitrogen atom combines with three *A* atoms to form a compound in which three pairs of electrons are shared in covalent bonds, and the outer electron shell of the nitrogen atom is filled.

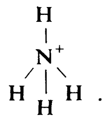
$$\cdot\overset{\cdot\cdot}{\underset{\cdot}{N}}\cdot + 3A\cdot \longrightarrow A{:}\overset{\cdot\cdot}{\underset{\underset{A}{\cdot\cdot}}{N}}{:}A$$

Shapes of nitrogen compounds

It has been shown by experiment and is found consistent with theoretical views that the orbitals (*i.e.,* the regions around the nucleus occupied by the four electron pairs) point to the four corners of a tetrahedron (triangular pyramid). Simple compounds in which three atoms share covalent bonds with nitrogen thus have the shape of a flattened pyramid, with the nitrogen atom at the apex and the unshared electron pair above it. In fact, however, this shape is very mobile and inverts (turns inside out like an umbrella) very rapidly. The interaction between the two forms is represented by an equation in which each bonded pair of electrons is a single line, and each single electron is a dot.

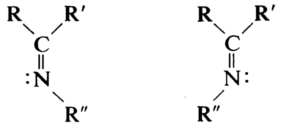


The unshared electron pair is available to form a fourth bond by sharing with another atom having an incomplete outer shell. The fourth bonding results in the formation of a tetrahedral molecule (or ion) with all four corners occupied by other atoms; an example is ammonium ion, NH_4^+, which can be represented:



Such species of bonded atoms maintain their form more or less rigidly when none of the four surrounding atoms is a hydrogen, and, if all four bonded atoms are different, the same kind of dissymmetry (left- and right-handedness, in reference to the position of the four relative to each other) as is encountered in carbon compounds may arise.

When a nitrogen atom shares two electron pairs with the same atom, a double bond results. The atoms joined by the double bond and the atoms attached directly to them then lie all in the same plane, together with the unshared electron pair on the nitrogen. Such compounds may have two geometrically isomeric forms (*i.e.,* forms differing only in the arrangement in space of the groups about a double bond); for example, many imines (compounds in which nitrogen is doubly bound to carbon) have isomers that may be represented as



The formation of a triple bond by sharing three electron pairs is most commonly encountered between a nitrogen and a carbon atom. Such compounds, called nitriles, have a linear structure (*i.e.,* the atoms lie in a straight line), as in the case of a single radical (*R*) bonded to a carbon atom that also shares a triple bond with a nitrogen atom, expressed as $R{-}C{\equiv}N$. (Molecular nitrogen and certain less common classes of compounds, the azides and diazonium salts, are considered to have triple bonds between two nitrogen atoms.)

CLASSIFICATION OF ORGANIC NITROGEN COMPOUNDS

Parent compounds

The parent inorganic compounds from which organic nitrogen compounds are obtained include ammonia (NH_3), hydrazine ($H_2N{-}NH_2$), hydroxylamine ($H_2N{-}OH$), hydrogen azide (HN_3), nitrous acid (HNO_2), nitric acid (HNO_3), and a few very unstable substances: diazene, or diimide, ($HN{=}NH$), nitroxyl (HNO), amino radical ($H_2N\cdot$), aminoxyl ($H_2N{-}O\cdot$), triazene ($H_2N{-}N{=}NH$).

Amines. Compounds derived from ammonia by replacement of one or more hydrogen atoms by hydrocarbon groups, which have varying numbers (indicated by *m* and *n*) of carbon and hydrogen atoms, C_mH_n, are known as amines; alkyl (*R*), or aryl (*Ar*). When only one hydrogen atom is replaced, the resulting compounds are called primary amines and are represented as RNH_2; when two hydrogens are replaced, they are secondary (R_2NH), and, when all three hydrogens are replaced, they are tertiary (R_3N). All such compounds resemble ammonia in that they are basic (*i.e.,* they form salts with acids by virtue of the unshared electron pair). As the hydrocarbon portion (the part consisting of groups containing only carbon and hydrogen atoms) of the molecule becomes larger, the amines become less volatile (*i.e.,* have less tendency to vaporize at ordinary temperature), less soluble in water, and generally more like hydrocarbons. Amines have lower melting points than most other classes of nitrogen compounds, and are more strongly basic. Most naturally occurring compounds of nitrogen are amines or derivatives of them.

Amines are named by prefixing the name of the attached groups to the stem amine, as in dimethylamine $(CH_3)_2NH$. If an aryl group such as the phenyl radical, $=C_6H_5$ (from benzene, C_6H_6), is attached, the compound is usually considered to be a derivative of aniline and is named as substituted aniline, as in 4-chloro-*N*-methylaniline ($ClC_6H_4NHCH_3$).

Derivatives of the ammonium (the "-ium" ending indicates, as in carbonium carbon, the presence of a localized positive charge) ion bearing four hydrocarbon groups (R_4N^+) are quaternary ammonium salts. The quaternary ammonium ions themselves are not basic (there being no longer an unshared electron pair), but their hydroxides (compounds in which the positive charge on the nitrogen atom is neutralized by the negative charge of a hydroxide ion, OH^-), $R_4N^+OH^-$, are strong bases that resemble sodium and potassium hydroxides. They are named by prefixing the names of the substituents to the stem ammonium, as in the tetramethylammonium ion ($[CH_3]_4N^+$).

Double bonds

Imines. Compounds in which nitrogen is bound to carbon by a double bond are known as imines (sometimes called Schiff bases). Imines are distinctly weaker bases than the amines. They are most commonly named by following the name of the related carbonyl compound by the word imine, as in diethyl ketone *N*-methylimine, $(C_2H_5)_2C{=}N{-}CH_3$, related to diethyl ketone, $(C_2H_5)_2C{=}O$.

Carbinolamines. Closely related to the imines, although containing no double bond, are the carbinolamines. They are produced by addition of the elements of water to the double bond and may usually be dehydrated to form imines (providing there is a hydrogen on the carbinolamine nitrogen atom). They are given names on the basis of being amines bearing a 1-hydroxy substituent (OH), as in 1-hydroxyethyldimethylamine,

$$CH_3CH(OH)N(CH_3)_2.$$

Triple bonds

Nitriles. The nitriles have a triple nitrogen-to-carbon bond (which is written $-C{\equiv}N$), generally have low melting points, and are for ordinary considerations totally nonbasic (do not neutralize acids to form salts). They have a close affinity with carboxylic acids—*i.e.,* organic acids containing the carboxylic group, $-COOH$, such as acetic acid, CH_3COOH, and benzoic acid, C_6H_5COOH; the carbons are in the same oxidation state, having in each case three bonds to an atom having higher affinity for electrons (oxygen or nitrogen). They are named by using the root name of the corresponding acid, with the suffix nitrile, as in acetonitrile, $CH_3C{\equiv}N$ (from acetic acid, CH_3COOH).

Amides. The replacement of an acidic hydroxyl group by an amino group produces the class of nitrogen compounds known as amides. The carboxamides, $RCONH_2$, are the most important group; other substituent groups may be attached to the nitrogen. Amides known as sulfonamides, RSO_2NH_2, are also produced from the sulfonic acids, RSO_3H. Most amides are solids, apt to have high melting points, and are essentially nonbasic and of low volatility. Most are named from the corresponding acid, using the suffix "-amide," as in acetamide, CH_3CONH_2, and *N*-ethylbenzenesulfonamide,

$$C_6H_5SO_2NHC_2H_5.$$

Imides. Compounds of the imide class have two carbonyl (*i.e.,* containing the organic radical $\diagdown\!\!\!\diagup C=O-$ *e.g.,* acetyl, CH_3CO-) groups attached to the same nitrogen, as in acetimide, $(CH_3CO)_2NH$. They are weakly acidic (*i.e.,* neutralize alkalies to form salts) if the nitrogen is not otherwise substituted. If both oxygens of a carboxyl group ($-COOH$, the radical characteristic of most organic acids) have been replaced by nitrogen, the amidine structure results. Amidines are moderately strong bases. The general formula is $RC(NH_2)=NH$.

Cyanic acid

Carbonic acid derivatives. The compounds derived from carbonic acid are usually considered as a class by themselves. The nitrile of carbonic acid is cyanic acid, $HO{-}C{\equiv}N$, which in turn can provide two groups of derivatives: the cyanates, $R{-}O{-}C{\equiv}N$, and the isocyanates, $R{-}N{=}C{=}O$. They are named by combining the class name with the name of the substituent, as in ethyl isocyanate, $C_2H_5{-}N{=}C{=}O$.

The amide of cyanic acid, cyanamide, $H_2N{-}C{\equiv}N$, and its tautomer (tautomers are compounds of the same chemical formula but different molecular structure that interconvert reversibly) carbodiimide, $HN{=}C{=}NH$, also yield a group of derivatives.

Carbonic acid, HOCOOH, which has two hydroxyl groups, reacts with ammonia to produce amides in two stages, carbamic acid, NH_2COOH, and urea,

$$NH_2CONH_2.$$

Carbamic acids are unstable in the free state but their salts and esters are well known. A salt is the compound other than water formed by reaction of an acid with a base, and an ester is a derivative of an acid in which a hydrocarbon group is attached to an oxygen atom in place of the acidic hydrogen atom. The esters are often called urethans, after the primitive name of the ethyl ester, $C_2H_5O{-}CO{-}NH_2$. Urea and its derivatives are almost invariably solids of high melting point and low volatility and are commonly named as substituted ureas, as *N,N′*-dimethylurea, $CH_3{-}NH{-}CO{-}NH{-}CH_3$. Thiourea is the compound in which a sulfur atom replaces the oxygen atom of urea.

The amidine of carbamic acid goes by the special name guanidine. The guanidines, $H_2NCNHNH_2$, are in general very strong bases, approaching hydroxide ion (OH^-) in strength.

Hydroxylamine compounds. Nearly all the various derivatives of ammonia (NH_3) have their parallel with hydroxylamine ($H_2N{-}OH$). Hydroxylamine may carry one, two, or three substituents, which may be bonded to either the nitrogen or the oxygen atom; *e.g.*, *N,N*-dimethylhydroxylamine, $(CH_3)_2N-OH$, and ethoxylamine, $C_2H_5-O-NH_2$. Such compounds are similar to amines but are weaker bases. Quaternary compounds corresponding to those of ammonia are of two types: tetrasubstituted salts and amine oxides.

The hydroxylamine compounds analogous to imines are the oximes, $RR'C{=}NOH$, and nitrones,

$$RR'C=\overset{+}{N}RO^-.$$

Nearly all are solids, but the oximes and derivatives have considerably lower melting points than the nitrones.

The most important hydroxylamine derivatives of the carboxyl group are the *N*-acyl compounds, called hydroxamic acids, $RC({=}O)NHOH$.

Hydrazine compounds. Hydrazine is also the parent of a family of structures analogous to the organic derivatives of ammonia. There are compounds of from one to four substituents (mono-, di-, tri-, and tetrasubstituted hydrazines); all are weakly basic. Quaternary hydrazinium compounds

$$R_2N-\overset{+}{N}R_3\,X^-$$

(in which X^- represents any negatively charged atom or group of atoms) are also well known. The analogues of imines, $R_2C{=}N-NR_2$, are called hydrazones and may have any of a range of organic groups bonded to the nitrogen atom. Azines, $R_2C{=}N-N{=}CR_2$, have two imine groups joined by an N—N bond. The only important hydrazine derivatives of the carboxyl group are the hydrazides, which are weakly basic.

Diazene compounds. The derivatives of diazene that have two substituted groups (*i.e.*, are disubstituted), $R-N{=}N-R$, are called azo compounds, such as azobenzene, $C_6H_5-N{=}N-C_6H_5$. They are neutral substances, generally coloured and solid. Monosubstituted diazenes are known but are very unstable.

The diazonium compounds,

$$Ar-\overset{+}{N}\equiv N\,X^- \rightleftharpoons Ar-N=N-X,$$

are closely related. The salts of strong acids are truly ionic (*i.e.*, electrically charged) and show the properties to be expected of a triple bond. They are generally rather unstable substances but of great importance as intermediates (chemicals produced as necessary steps between starting material and finished product) in synthesis. The diazo compounds, such as diazomethane,

$$CH_2=\overset{+}{N}=\overset{-}{N} \rightleftharpoons \overset{-}{C}H_2-\overset{+}{N}\equiv N,$$

also show evidence of a triple bond.

Compounds with three or more nitrogens. Triazenes and azides are among the types of compounds having chains of three or more nitrogens. The word triazene denotes three nitrogen atoms and a double bond, $H_2N-N{=}NH$, and, in analogous fashion, names for longer chains and different degrees of unsaturation (that is, capable of combining directly with additional groups of atoms) may be developed.

Azides of various types, such as methyl azide, CH_3-N_3, phenyl azide, $C_6H_5-N_3$, and acetyl azide, CH_3CO-N_3, all contain the linear (*i.e.*, all the atoms lie in a straight line) azido group

$$-N=\overset{+}{N}=\overset{-}{N} \rightleftharpoons -\overset{-}{N}-\overset{+}{N}\equiv N.$$

The nitroso and nitro compounds, $R-N{=}O$ and $R-NO_2$, represent higher oxidation states. Both are neutral, and the former are all a beautiful blue or green, whereas the latter are colourless to yellow. Most nitroso compounds, such as nitrosobenzene, $C_6H_5-N{=}O$, are in equilibrium with colourless dimers (a dimer is a molecule formed by the union of two simpler molecules), with the nitrogen molecules bonded together. Both classes of compound are invariably named by means of the "nitroso-" or "nitro-" prefix.

Nitrous and nitric acid compounds. The highest oxidation states of nitrogen are represented by the derivatives of nitrous acid (HNO_2) and nitric acid (HNO_3). Among them are the esters, $R-O-N{=}O$ and $R-O-NO_2$, which contain no carbon-to-nitrogen bond. The nitrites that are exceptionally volatile, and the nitrates, such as nitroglycerin (or glyceryl trinitrate), are potentially explosive in proportion as the ratio of oxygen to carbon and hydrogen in the compound approaches the optimum for complete internal combustion to water, oxides of carbon, and molecular nitrogen. Nitrous and nitric acids can give rise to amides, hydroxamic acids, and hydrazides, just as do other acids. All of them are known in the form of organic derivatives.

The volatility of nitrites and nitrates

THE AMINES AND THEIR DERIVATIVES

Physical and chemical properties. The smallest amines are gases, but most amines are liquids, and all of them except the very large have fishy or musty odours. The volatility (*i.e.*, the tendency to vaporize at ordinary temperatures) decreases with increasing molecular weight and increasing number of hydrogens attached to the nitrogen. Primary amines thus have the highest boiling points and tertiary the lowest for comparable molecular weights. The smaller amines are soluble or even completely miscible in water (that is, can be mixed completely without separating into two layers) but, when the size of the molecule exceeds five carbon atoms, the solubility drops rapidly. Nearly all amines are soluble in the common organic solvents. Aliphatic amines (*i.e.*, those in which the nitrogen atom is not attached to a benzene ring) are slightly stronger bases than ammonia and will turn litmus indicator paper blue, but aromatic amines (those in which the nitrogen atom is attached to a carbon atom present in a benzene ring) are considerably weaker, and most of them will not affect litmus.

Aliphatic amines

Salts. Nearly all salts of amines are nonvolatile solids, very soluble in water, to which they give a weak acidic reaction if they are salts of strong acids, such as hydrochloric, nitric, or sulfuric; generally they do not dissolve in nonpolar solvents such as ether and the hydrocarbons. Quaternary ammonium compounds are completely ionic in nature—the hydroxides are alkalies as strong as sodium hydroxide (lye).

Sources and preparation. Some simple aliphatic amines occur naturally, mostly as the products of decay, but there are no practical natural sources. Industrially, aliphatic amines are mostly made from fatty acids (organic acids derived from natural fats and oils), olefins (unsaturated aliphatic compounds whose molecules contain only carbon and hydrogen atoms; *i.e.*, they are hydrocarbons), or alcohols (alcohols are hydroxyl compounds, containing the oxygen–hydrogen radical, OH, bonded to a saturated carbon). The fatty acids (or the fats and oils in which they occur bound) may be reduced to alcohols (one important step in production of detergents), which can then be converted to amines catalytically by reaction with ammonia or by indirect, laboratory methods involving initial conversion to a suitably reactive derivative. Alternatively, the fatty acids may be converted to nitriles (by way of their amides) by reaction with ammonia, followed by hydrogenation (combination of hydrogen with another substance, usually an unsaturated organic compound). These reactions also give rise to secondary and tertiary amines.

A few amines, notably aniline and some of its derivatives, are made by reduction (reduction is an increase in the number of electrons) of nitro compounds, although in recent years it has become more common to produce aniline from chlorobenzene in a continuous process under high pressure.

Reductive alkylation (a process of adding alkyl radicals and hydrogen), whereby an aldehyde or ketone is treated with ammonia and hydrogen in the presence of a catalyst, has both industrial and laboratory utility.

There are three types of reaction by which amines may be prepared on a laboratory scale: displacements on halides or derivatives; reduction of nitriles, amides, oximes, nitro compounds, etc.; and rearrangements (Hofmann, Curtius, Lossen, Beckmann, Schmidt). Many variations of them are described in specialized works.

Chemical behaviour of amines

The chemical behaviour of amines is dominated by their basicity. Nearly all of them form salts even with moderately weak acids, and the aliphatic amines even extract carbon dioxide from moist air to form salts. Since salts in general are soluble in water, most amines will thus dissolve in dilute solutions of acids. Salts also generally crystallize well, and, by reaction with acids, liquid or gaseous amines can be converted to solid derivatives useful for identification.

Oxidation characteristics. Aliphatic amines are not easily oxidized (oxidation is loss of electrons), but aniline and its derivatives commonly become coloured slowly on prolonged contact with air, owing to oxidation of the activated benzene ring to form complex dyes. Since amines are the lowest oxidation state of nitrogen, they are not ordinarily reducible (except for hydrogenolysis of benzylamines and some quaternary ammonium salts, in which a C—N bond is broken with addition of hydrogen to each fragment).

Chlorine and bromine replace the *N*-hydrogens of amines to form *N*-haloamines. Tertiary amines undergo more deep-seated changes.

Nitrogen gas formation

The reaction of amines with nitrous acid is an old and important reaction. It appears that all primary amines give a diazonium salt, $R—N_2^+X^-$, but only with aniline and its derivatives can such products be isolated (or even detected). The observed result with aliphatic primary amines is the formation of mixtures, consisting of olefins, cyclopropanes, and isomeric alcohols, with evolution of nitrogen. The formation of nitrogen gas (effervescence) has been used as a qualitative test for primary amines. From secondary amines, nitrosamines precipitate as nonbasic, yellowish oils, and tertiary amines are degraded to mixtures of nitrosamines with aldehydes or ketones derived from one of the alkyl groups.

Aromatic primary amines are a special case. Cold (about 0° C), they react with dilute nitrous acid to form solutions of diazonium salts, which easily lose nitrogen on warming and form phenols (*Ar*OH).

Alkyl halides, sulfates, and related alkylating agents react more or less readily with free amines so as to add an alkyl group to the nitrogen atom. The process may be repetitive, especially if an inorganic base is present, so that most amines are eventually converted to quaternary salts, if bulk does not interfere.

Reactions with other organic compounds. The reactions of amines with other classes of organic compounds form a vast subject. In the most general terms, they do not react with alcohols or olefins except at high temperatures in the presence of surface catalysts, when they may become alkylated on the nitrogen. Aldehydes and ketones will usually react with primary amines to form imines by addition at the carbonyl group followed by elimination of water, as shown in the equation

$$R-NH_2 + R'_2C=O \longrightarrow R'_2C=NR + H_2O.$$

Carboxylic acids form salts, which may lose water when heated very strongly, forming amides. Primary and secondary amines react to form amides, vigorously with acid chlorides, slightly less vigorously with acid anhydrides, and extremely sluggishly with esters.

$$RCOX + 2R'NH_2 \rightarrow RCONHR' + R'NH_3^+X^-$$

Amines do not react with other bases, weak or strong, except insofar as amines are liberated from their salts. Quaternary ammonium salts, however, undergo the Hofmann reaction when they are converted to the hydroxides (usually by treating them with moist silver oxide) and then heated strongly. One group is cleaved from the nitrogen as an olefin leaving a tertiary amine.

$$R_3\overset{+}{N}-CH_2CH_2R\,OH^- \xrightarrow{\Delta} R_3N + CH_2=CHR + H_2O$$

Compounds with amide linkages. *Physical properties.* The carboxamides and imides are mostly solids of low volatility and thus little odour, although they are likely to have flavour. Unsubstituted amides, RCONH$_2$, have the highest melting and boiling points; substitution of hydrogen by alkyl groups (paraffin hydrocarbon radicals) reduces the intermolecular attraction by removing the capability for hydrogen bonding between N—H of one molecule and C═O of another. The small amides and imides (less than six carbon atoms) are soluble in water, and formamide is completely miscible.

Common characteristics of amides

Similar remarks can be made about most of the other types of amides: thionamides, sulfonamides, ureas, urethans (carbamates), amidines, and imidates (imino esters). The last two types, however, which have no carbonyl group, have lower melting points, and the simple imidates are liquid at room temperature.

Amides in general are neutral in water solution, but they show a feeble basicity toward very strong acids, especially in the absence of water. Imidate esters, however, are almost as basic as alkylamines, and amidines are the strongest organic bases short of the quaternary ammonium hydroxides. Sulfonamides show no basic properties, but those having at least one hydrogen on the nitrogen have acidities similar to those of phenols.

Sources and preparation. There are no practical natural sources of simple amides, although polyamides occur in great abundance as the protein of living systems. Simple amides are made industrially from ammonium salts of carboxylic acids by strong heating; some *N*-substituted amides are made analogously. Esters react with ammonia to form amides, but the general laboratory method is to treat an amine with an acid chloride or anhydride. It is usually desirable to add an equivalent of alkali (base, usually having a hydroxyl group that ionizes to hydroxide ion, OH$^-$, which then combines with the hydrogen ion, H$^+$, of an acid to form water, a process called neutralization) to neutralize the acid that would otherwise bind a second equivalent of amine (Schotten–Baumann procedure).

Imides do not occur naturally except for some complex heterocyclic compounds. For preparative purposes, imides can generally be prepared from amides by treating them with acid chlorides or anhydrides or with metallic sodium or potassium, but cyclic imides, derived from dicarboxylic acids, can in many instances be prepared by strongly heating the ammonium salt of the acid or the half-amide. Imides can also be prepared by heating nitriles with carboxylic acids.

The derivatives of carbonic acid are a special case. Urea is prepared commercially by heating ammonia with carbon dioxide, but substituted ureas are best prepared by the reaction of an isocyanate with an amine. Urethans are also made from isocyanates, by reaction with alcohols.

The most characteristic reaction of amides is hydrolysis (a chemical reaction with water), by which they are converted to an acid and an amine. Amides are generally inert to pure water and require a mineral acid, alkali, or enzyme to catalyze hydrolysis. Wool and nylon, both of which are polyamides, can thus be dissolved by heating them with strong acid or base such as potassium hydroxide (KOH). The equation is

$$RCONR_2 + KOH \longrightarrow RCO_2K + R_2NH.$$

Amides can also be dehydrated, most commonly by heating with phosphorus pentoxide or phosphoryl chloride. Unsubstituted amides give nitriles, and monosubstituted amides give ketenimines, but disubstituted amides are inert. Monosubstituted formamides can be dehydrated to isocyanides.

Amides are not readily oxidized and are inert to most reducing agents. Catalytic hydrogenation will convert amides of aliphatic acids to amines at somewhat high pressures and temperatures, and electrolytic reduction is generally successful with aromatic amides, but the most generally effective reagent is lithium aluminum hydride. Conditions may generally be chosen so as to favour either amines, alcohols, or aldehydes as the products.

Aldehydes but not ketones generally react with amides that have at least one hydrogen on the nitrogen. The first step is addition to the aldehyde carbonyl group, but this may be followed by reaction with a second molecule of amide and elimination of water. The technique of preserving flesh by immersion in formaldehyde solution makes use of these reactions, and the condensation of formaldehyde with urea (carbamide) produces the industrially important urea–formaldehyde resins.

$$H_2NCONH_2 + H_2C=O$$
$$\longrightarrow H_2NCONHCH_2OH + (-NHCONHCH_2-)_n$$

Amides may react with acid chlorides or anhydrides, especially if a base is present, to form imides.

The halogen elements and the hypohalous acids, HO*X*, replace a hydrogen on the nitrogen in the presence of base. The resulting *N*-haloamides are not very stable and easily undergo the Hofmann rearrangement when warmed with excess base. The overall effect is to convert the amide to a primary amine having one fewer carbon atom.

Amides suffer replacement of the amino group by an aryl or alkyl group, to form ketones (formamides yield aldehydes) in two ways: reaction with Grignard reagents or treatment with an aromatic hydrocarbon, such as benzene, in the presence of phosphoryl chloride (Vilsmeier–Haack reaction).

Compounds with double and triple C-N bonds. *Imines, nitriles, isocyanates, etc.* The simplest members of each class in this section are liquids, although little more can be said about the small imines, for they polymerize so readily. Hydrogen cyanide is actually the smallest nitrile. Acetonitrile has an unusually high dielectric constant and is a polar solvent capable of dissolving many salts as well as most other organic compounds. The isomer methyl isocyanide is a liquid of exceptionally revolting stench; methyl isocyanate has a pungent, biting odour; and methyl isothiocyanate smells somewhat savory as well as pungent. Cyanamide forms transparent crystals. The isomeric (forming isomers) carbodiimides polymerize too rapidly for the simpler ones to have been obtained pure, but in general they are liquids similar to the cyanamides.

The small imines, nitriles, cyanamides, and carbodiimides are miscible or soluble in water, but isocyanides, isocyanates, isothiocyanates, cyanates, and thiocyanates are nearly insoluble.

Imines are most generally prepared either by reaction of a ketone or aldehyde with a primary amine or by the addition of Grignard reagents to nitriles. Oxidation (or catalytic dehydrogenation) of amines can also be made to give imines.

Preparation of isocyanates

Isocyanates are nearly always prepared by the reaction of primary amines with phosgene or by a process called, after its discoverer, the Curtius rearrangement of acyl azides, although other methods exist, such as the addition of cyanic acid to olefins, the reaction of potassium cyanate with alkylating agents, such as methyl sulfate, and the cleavage of certain urethans by heat or salts of silver or mercury. Isothiocyanates, *R*—N═C═S, can be obtained naturally in some cases; allyl isothiocyanate, for example, comes from mustard seeds, and an old term for isothiocyanates as a class is mustard oils. Laboratory preparation starts with primary amines, which react with carbon disulfide to form dithiocarbamate salts. These can be decomposed to isothiocyanates, sometimes by heating but more generally by treating with salts of silver or of lead. An important source of thiocyanates is the reaction of alkyl thiosulfates with alkali metal cyanides.

Sources of nitriles

The commonest source of nitriles is the reaction of alkyl halides with sodium (or potassium) cyanide, which goes well with primary alkyl halides but fails when they are tertiary. When silver cyanide is used, a complex is formed, which liberates an isocyanide when broken up with excess potassium cyanide. Aryl halides, which are generally inert toward alkali metal cyanides, can be converted to cyanobenzenes by heating with cuprous cyanide. Both nitriles and isocyanides can be made by dehydrating amides.

An industrially important source of nitriles is the catalytic reaction of hydrocarbons bearing methyl groups with ammonia and oxygen at high temperatures (ammoxidation).

Carbodiimides are prepared from thioureas by treating with reagents that extract hydrogen sulfide (*e.g.,* mercuric oxide), as well as by other methods.

Principal reactions. The double and the triple C—N bonds are susceptible to hydrolysis (reaction with water), generally with catalysis by either acid or base (a catalyst alters the speed of reaction enormously, without being itself used up, generally at a much lower temperature than needed to initiate the reaction without a catalyst), with varying degrees of ease. The ultimate result is replacement of all C—N bonds by C—O bonds, such that imines yield aldehydes or ketones, nitriles yield carboxylic acids, isocyanides yield primary amines and formic acid, and cyanates, isocyanates, and their sulfur analogues yield carbon dioxide (or carbon oxysulfide). When conditions are mild enough or the amount of water is limited, hydrolysis may be stopped at the intermediate stage of amides (from nitriles), formamides (from isocyanides), or ureas (from isocyanates and carbodiimides).

Reactions with alcohols

Hydrolysis is paralleled by reactions with alcohols, of which the two most important examples are the conversion of nitriles to imidate esters and to ortho esters and of isocyanates to urethans. Ammonia and amines react analogously, particularly with isocyanates or isothiocyanates, and are thereby converted to ureas or thioureas, respectively.

Reduction (a category of reactions, the reverse of oxidation and here usually meaning reaction with hydrogen in the presence of a catalyst) in general saturates the double and triple C—N bonds with hydrogen, converting the compounds to amines of one kind or another.

A class of compounds called Grignard reagents reacts readily with nitriles and isocyanates, converting them to imines and amides, respectively. The same conversion can be accomplished by reaction with aromatic hydrocarbons and a class of catalysts called Friedel–Crafts catalysts; *e.g.,* aluminum chloride.

COMPOUNDS WITH N—N BONDS

Physical properties. The smaller hydrazines are hygroscopic liquids with unpleasant odours and are nearly as strongly basic as amines when they bear only one or two alkyl substituents. Successive substitution lowers the basicity, and the tetraalkylhydrazines, R_2N-NR_2, are only feebly basic. Substitution of methyl groups for hydrogen in hydrazine lowers the boiling point, by reducing hydrogen bonding, but larger substituents add so much weight that their net effect is to raise the boiling point.

Substitution of an acyl group for hydrogen, giving a hydrazide, raises both the melting and boiling point, and such compounds are nearly all solids, basic enough to dissolve in dilute mineral acid but weaker than hydrazine. Hydrazones are intermediate in properties and generally are liquids or low-melting solids (unless there are large substituents on the nitrogen atom) and of feeble basicity. Azines, $R_2C=N-N=CR_2$, may be liquids or solids and are generally yellowish and nonbasic.

The derivatives of azobenzene, a typical azo compound, in general are solids, always coloured, insoluble in water, and nonbasic; the larger aliphatic azo compounds are liquids. Monosubstituted analogues, usually named as diazenes, are of low stability, and much less is known about them.

Azoxy compounds resemble the azo compounds but are of paler colour; they are distinct from the oxadiaziridines, with which they were at one time confused.

Diazonium salts are generally colourless, crystalline, water-soluble substances that decompose slowly at room temperature and are capable of exploding. The salts are neutral in reaction, which implies that the diazonium hydroxides are strong bases. In basic solution, however, diazonium salts become converted to diazotate ions, $Ar-N=N-O^-$, which can exist in stereoisomeric forms related to those of the imines illustrated previously. Closely related are the diazo compounds, of which diazomethane, $CH_2=N_2$, is the best known; they are coloured yellow to purple, generally low-melting, and seldom very stable; they are inert to bases but are destroyed by acids.

The azides are colourless compounds, commonly liquid and insoluble in water. They are neutral but may be destroyed by concentrated acids, and though they are more stable than diazo compounds they may be dangerously explosive, especially if they are small.

The most important of the remaining compounds with N—N bonds are the triazenes, $R_2N-N=N-R$, but all are rarely encountered.

Preparative methods

Sources and preparation. Alkyl hydrazines are produced when hydrazine is treated with alkylating agents, and, although mixtures with dialkyl and even trialkyl hydrazines may result, this is usually a satisfactory preparative method. Some alkyl hydrazines are more easily prepared from hydrazides by reduction with lithium aluminum hydride. Reduction of nitrosamines is a good route to *N,N*-disubstituted hydrazines, and reduction of hydrazones has been used to prepare trisubstituted hydrazines. Aryl hydrazines can be obtained in wide variety by reducing diazonium salts; although this is the preferred method, a good alternative is the reaction of azodicarboxylic ester with aromatic hydrocarbons. This reaction produces hydrazine analogues of urethans, which can be hydrolyzed cleanly to aryl hydrazines.

Hydrazones are usually made by the reaction of ketones or aldehydes with hydrazines, although they can also be made by tautomerization of azo compounds. Hydrazides are made by reaction of esters with hydrazine; if acid chlorides are used, diacyl hydrazines are produced.

Aromatic azo compounds can be prepared in great variety by the reaction of diazonium salts with compounds containing reactive benzene rings; this is the method used to prepare azo dyes. Alternatively, an aromatic primary amine and a nitroso compound can be converted to an azo compound by the elimination of water. When the desired azo compound is symmetrical, such as azobenzene, the same effect can be obtained by reducing the corresponding nitrobenzene under suitable conditions. Aliphatic azo compounds are nearly always prepared by oxidizing a hydrazine.

Azoxy compounds are formed by the oxidation of azo compounds with peroxides or by reaction of nitroso compounds with hydroxylamines, with elimination of water. Reduction of a nitrobenzene may be controlled so as to produce the latter two reagents, thereby producing an azoxybenzene.

Diazonium compounds are always prepared by diazotization, the reaction of a primary aromatic amine with cold nitrous acid (although there are some unimportant reactions that also give rise to them).

The most general method among many for preparing diazoalkanes is the treatment of an *N*-nitroso-*N*-alkylamide with a strong base. Oxidation of simple hydrazones is a useful preparative method for larger diazoalkanes.

Aliphatic azides are usually prepared by a simple displacement reaction between an alkyl halide and sodium azide, but aryl azides, which cannot in general be prepared in this way, are easily prepared from diazonium salts and sodium azide. Both aryl and acyl azides can be prepared by the reaction of nitrous acid with the corresponding hydrazine derivative.

Reactions. The most characteristic reactions of alkyl and aryl hydrazines are salt formation and alkylation. Alkyl halides, sulfates, etc., will replace the hydrogens stepwise, usually continuing at the same nitrogen until a quaternary compound is formed. Hydrazones and hydrazides undergo the same type of reaction, although not as readily.

Most organic hydrazine derivatives are reducing agents in some degree, and those that bear no more than one substituent on each nitrogen are especially easily oxidized, forming azo compounds. With two hydrogens on the same nitrogen, oxidation is a little more difficult but can be accomplished with strong oxidizing agents so as to convert disubstituted hydrazines to diazenium salts and hydrazones to diazoalkanes. The N—N bond in hydrazines can be cleaved by reduction to give two molecules of amine; the preferred reagent is hydrogen in the presence of a nickel catalyst.

Nitrous acid attacks most hydrazine derivatives; the initial product in all cases is probably a nitroso (*i.e.,* the molecule contains the —NO group) derivative, but only when the hydrazine is trisubstituted can it usually be isolated. Monoarylhydrazines and simple hydrazides are converted all the way to azides, in an important preparative reaction.

Azo compounds are relatively unreactive. They are inert to alkylating agents, to aldehydes and ketones, and to acylating agents, and are not easily attacked by strong acids or bases. Reduction, their most characteristic reaction, takes place in two stages, first to a hydrazine and then to a pair of amines. Azo compounds having a hydrogen on an adjacent carbon isomerize (*i.e.,* rearrange their molecular structure) rather easily to hydrazones, especially in the presence of a base.

Reaction with acids

The most important general reaction of diazoalkanes is that with acids, which are converted to their esters, usually quantitatively and rapidly. Another important general reaction is the loss of nitrogen to form transient carbenes (electron-deficient species with divalent carbon) when heated or exposed to light. Diazomethane and to a lesser extent other diazoalkanes react with acid chlorides to form diazoalkyl ketones. These undergo the Wolff rearrangement when heated (usually in the presence of a silver catalyst) and form ketenes (or, in the presence of alcohols, water, or ammonia, the corresponding esters, acids, or amides). The overall process thus converts a carboxylic acid through its acid chloride to an acid with one more carbon atom (Arndt–Eistert synthesis).

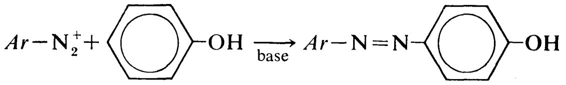
The most important reactions of azides are decomposition by heat, light, or concentrated acids, reduction, and addition of certain very strong bases. Heat or light causes loss of two-thirds of the nitrogen as N_2. In some cases this is accompanied by rearrangement, as in the Curtius rearrangement of acyl azides to isocyanates. In others, a reactive, electron-deficient intermediate, called a nitrene, is produced, which cannot be isolated but reacts very rapidly with itself or the solvent or both.

Many reducing agents, such as hydrogen (with a platinum catalyst) or sodium arsenite, convert azides cleanly to primary amines, for which this process is sometimes a useful preparative method.

The reactions of diazonium salts have been studied intensively since the mid-19th century, and the possibilities are so many and complex that only a small selection can be mentioned here. One important family of reactions involves loss of nitrogen and replacement of the diazonium grouping by another atom or group of atoms. Warming in water solution brings about replacement by OH, giving a phenol. In the presence of inorganic iodides or azides, aryl iodides or azides form rapidly. Replacement by chloride, bromide, or cyanide is brought about with the help of a copper(I) salt as catalyst (Sandmeyer reaction). Replacement by hydrogen may be brought about by alcohol or hypophosphorous acid.

Formation of azo compounds

Another important family of reactions consists of the attachment of another group, usually through a carbon atom, at the end of the diazonium system, to give various types of azo compounds. The most important examples involve reaction with phenols or anilines, forming substituted azobenzenes. When these substances also bear functions that help in binding them to textile fibres, they are azo dyes.



COMPOUNDS WITH N—O BONDS

Classes, sources, and preparation. Hydroxylamines are liquids or low-melting solids, volatile if small, and weakly basic. Oximes are still weaker bases and feeble acids and hydroxamic acids are weak acids similar to phenols. Nearly all hydroxylamine derivatives are soluble in water if they do not contain more than six carbon atoms.

Compounds with a nitroso group attached to carbon are bright blue or green, low-melting, neutral substances, but few of them are well known, owing to their capacity for rapid dimerization (usually reversible). The dimers are colourless solids having structures analogous to azoxy compounds. With many nitrosobenzenes, the colourless dimer dissociates to the green monomer simply on melting or dissolving in an inert solvent.

Most simple nitro compounds are liquids and are only slightly soluble in water, if at all. Many nitrobenzene derivatives are yellow. Nitro compounds are in general neutral, but those that bear a hydrogen on the same carbon as the nitro group are pseudoacids; that is, although only feebly acidic themselves, they react slowly with bases to form salts derived from tautomeric forms (nitronic acids, or *aci*-nitro compounds), which have acid strengths similar to those of carboxylic acids. Trinitromethane, however, exists entirely in the nitronic acid form and is a strong acid. Nitro compounds are polar, and the smaller nitroalkanes dissolve many salts.

The esters of nitrous and nitric acids are liquids, volatile if small, and insoluble in water. The nitrites are not very stable to storage and are also easily hydrolyzed; the nitrates are considerably more stable, if they are pure, and are difficult to hydrolyze. Nitrate esters are sensitive to heat and shock, especially when impure, and can be violently explosive.

Stable organic derivatives, known as nitrosamines, of the unstable amide of nitrous acid are well known. There is no practical natural source of compounds containing N—O bonds, although some of them occur in nature.

The preparation of organic derivatives of hydroxylamine relies heavily on hydroxylamine itself, which reacts with aldehydes and ketones to form oximes and with esters, anhydrides, or acid chlorides to form hydroxamic acids.

Nitroso compounds are most generally made by oxidation of *N*-substituted hydroxylamines, although many nitrosobenzenes can be made conveniently by oxidizing the corresponding aniline with peroxysulfuric acid, and a few can be made by direct nitrosation (addition of nitroso group, NO) of the benzene ring by nitrous acid (*e.g.*, *p*-nitrosodimethylaniline).

Preparation of nitro compounds

Nitro compounds are prepared in great variety by treating aromatic hydrocarbons with nitric and sulfuric acids; with some severe limitations, nitration of aliphatic hydrocarbons (generally at high temperature in the vapour phase) is also of preparative value, but most aliphatic nitro compounds are made by further reactions of the few commercially available nitroalkanes, such as nitromethane.

Esters of nitrous acid are easily made by reaction between dilute nitrous acid and alcohols, but a mixture of nitric and sulfuric acids is generally required to convert alcohols to alkyl nitrates. Nitrosamines are formed easily when secondary amines are brought into contact with nitrous acid, but nitramines require special methods, such as treating primary amines with *a*-cyanoisopropyl nitrate or converting an amine to a carbamyl chloride and then treating it with silver nitrate.

Reactions. Although most hydroxylamines are indefinitely stable as their salts, the free bases do not generally store well; *N*-substituted hydroxylamines slowly transform into amines and azoxy compounds, but *O*-alkyl hydroxylamines are more stable. *N*-aryl hydroxylamines differ from the alkyl analogues in that acid easily brings about a rearrangement to form *p*-aminophenols and other ring-substituted anilines.

Both alkylation with alkyl halides and acylation with esters, anhydrides, and acid chlorides take place easily. Aldehydes and ketones react with *O*-substituted hydroxylamines to form oxime derivatives, but *N*-substituted hydroxylamines generally react only with aldehydes, with which they form nitrones.

Hydroxylamines and oximes can be reduced by various reagents, most generally by catalytic hydrogenation, and then form amines.

Hydroxamic acids are hydrolyzed to the parent carboxylic acids by hot aqueous mineral acid. In the form of their alkali metal salts, they can be alkylated or acylated.

The chemistry of nitroso compounds is dominated by their oxidation and reduction. They are active oxidizing reagents, and a wide variety of reducing agents will convert them first to hydroxylamines and then to amines. Because nitroso compounds react with both of these functional groups, the actual products obtained from a reduction may be further transformation products—azo or azoxy compounds, hydrazines, benzidines, etc. Oxidation of nitroso to nitro compounds requires a strong oxidizing agent, such as permanganate or a peroxy acid.

Aliphatic nitroso compounds are generally so tightly bound up as the dimers that reactions specifically of the monomer are not easily observed. If there is a hydrogen on the same carbon as the nitroso group, the monomer easily and irreversibly isomerizes to an oxime. The nitroso group is an effective trap for free radicals, which add to the nitrogen to form a nitroxide, $RR'\text{N—O}\cdot$. Nitroso compounds react with many other types of compounds under suitable conditions: olefins, aldehydes, Grignard reagents, compounds having active methylene groups, etc.

Reduction of nitro compounds

By far the most important reaction of nitro compounds is their reduction, which, although not so easily accomplished as with nitroso compounds, can still be brought about by a large variety of reducing agents; hydrogen (with catalysts), dissolving metals (iron, tin, zinc, etc.), sulfides, sodium alkoxides, etc. The products may be any of the oxidation states of nitrogen but generally not nitroso compounds, for these are invariably reduced further. If the reducing medium is acidic, reduction almost always goes all the way to an amine. In neutral medium, it is in many cases possible to stop at the hydroxylamine stage. In basic solution, N—N bonds are usually formed, and the products are hydrazo, azo, or azoxy compounds.

The more important part of the chemistry of aliphatic nitro compounds has to do with the reactivity of the hydrogens on the same carbon as the nitro group. Concentrated sulfuric acid isomerizes primary nitroalkanes to hydroxamic acids, which are usually hydrolyzed to carboxylic acids and hydroxylammonium sulfate. Their salts, the nitronates, are not very stable and undergo some unusual reactions. Strong mineral acid cleaves them to aldehydes or ketones and nitrous oxide (Nef reaction). Strong base causes more complex reactions; primary nitroalkanes condense to heterocyclic compounds (isoxazoles) and nitromethane itself is converted in stages to a salt of nitroacetic acid. Heavy-metal salts have distinctive reactions, such as that of the mercuric salt of nitromethane, which loses water to form mercuric fulminate.

Most nitroalkanes decompose above 300° C (572° F) and form olefins, water, and oxides of nitrogen. With aldehydes (and some ketones), primary nitroalkanes react in a way analogous to the base-catalyzed aldol condensation, adding to the carbonyl group to form a nitro alcohol or its dehydration product, a nitro olefin (Henry reaction).

Alkyl nitrites decompose rather easily under the influence of heat or light to produce mixtures of carbonyl and *C*-nitroso compounds or their further transformation products. The first step is apparently loss of NO to form an alkoxy free radical (*i.e.*, a group of atoms with one unpaired electron), $R\text{—O}\cdot$. Alkyl nitrites also react as nitrosating agents, similar to nitrous acid.

Alkyl nitrates begin to decompose near 150° C (302° F). If the heating is carefully controlled, alkyl nitrites may be formed, along with smaller fragments and oxides of carbon, but otherwise explosion ensues, and the entire molecule may be converted to oxides of carbon, water, and nitrogen. Nitrates are most conveniently converted to the parent alcohol by reduction (zinc and acetic acid, ferrous chloride, etc.), which at the same time converts the nitrogen to nitric oxide or ammonia. Alkyl nitrates can function as alkylating agents or nitrating agents.

(P.A.S.S./Ed.)

Organic phosphorus compounds

Organic phosphorus compounds are carbon-containing compounds that also contain one or more atoms of the element phosphorus. Of the several million known organic compounds, several hundred thousand contain phosphorus.

The organic compounds of phosphorus are typically colourless liquids or solids, not physically distinct from other organic compounds of similar molecular size and character. Several types are very toxic, including the phosphines, but they present little danger in normal use because their exceedingly offensive odours generally prevent overexposure. Many toxic phosphorus compounds of other classes, however, are particularly hazardous because they are colourless, almost odourless liquids. Some of them are used as insecticides. Others were developed during World War II as a particularly unpleasant class of chemical warfare agent, the nerve gases.

Insecticides and nerve gases belong to the general fam-

Organic phosphates

ily of organic phosphates, which are probably the most important of the organic phosphorus compounds. Also included in this category are many biochemical compounds essential to life processes, such as the nucleic acids (control factors in heredity) and nucleotide coenzymes (compounds that permit enzymes to function properly). Other organic phosphates find industrial uses—as solvents and flame-retardation agents, for example. Other classes of organic phosphorus compounds are less useful, but some play important roles in the synthesis of complex chemical compounds of various types, such as pharmaceuticals.

There are many problems connected with the naming of organic phosphorus compounds, largely because the first organic phosphorus compounds were made and named in the early part of the 19th century, before complete knowledge of their structures was obtained. Furthermore, different authors in different countries have frequently used different names for the same compound. This situation was improved, however, after an international agreement on nomenclature was reached in 1952. In this article standard international nomenclature is used.

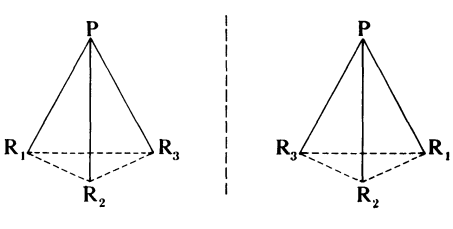
GENERAL PROPERTIES

The chemistry of the organic phosphorus compounds is best understood in terms of the influence of the phosphorus atom on the organic molecule. This influence depends in large measure on the electronic character of that atom.

Electronic configuration. Atoms of phosphorus, like those of other elements, contain a central, positively charged nucleus and, surrounding it, successive concentric rings, or shells, of negatively charged electrons. The electrons of the outermost shell are capable of joining with the electrons of other atoms to form chemical bonds, and they are, therefore, called bonding, or valence, electrons—the term valence referring to the number of other atoms with which a single atom of a given element can combine. In its outermost shell, a phosphorus atom has five electrons in two subshells called—for no important reason—*s* and *p* subshells. Within these subshells the electrons move in particular paths, or orbitals. The five outermost, or valence, electrons of a phosphorus atom occupy one *s* orbital (containing two electrons) and three *p* orbitals. The three *p* electrons are unpaired (each orbital being capable of holding two electrons) and therefore are available to form three single bonds, giving rise to trivalent compounds of phosphorus. Alternatively, however, the two *s* electrons can also become involved in bonding, and this situation leads ultimately to pentavalent phosphorus compounds.

Pyramidal structures

Structure and bonding. Trivalent phosphorus compounds, with the general formula PX_3 (in which P represents a phosphorus atom and X any other atom or group), have pyramidal structures (one atom at each vertex of a three cornered pyramid), that are closely analogous to those of the corresponding and better known nitrogen compounds. One of the main differences between the phosphorus and nitrogen compounds is that the angles between the different $P-X$ bonds are considerably smaller than the corresponding angles between the $N-X$ bonds (in which N is the symbol for nitrogen)—close, in fact, to the 90° expected if bonding involves only the *p* orbitals of phosphorus. Another difference between phosphorus and nitrogen compounds is that the pyramidal configuration in the phosphorus series is much more stable to inversion—that is, to turning inside out, like an umbrella. As a result, trivalent organic phosphorus compounds (phosphines) with three different organic substituents, unlike the corresponding nitrogen compounds (amines), can exist in two separate forms (isomers), which are stable under mild conditions. The two forms are mirror images of one another, as shown by the diagrams below:



In these diagrams, R_1, R_2, and R_3 represent the different organic groups joined to a phosphorus atom. The solid lines represent bonds between the phosphorus atom and the organic groups (and also the edges of the pyramidal structures of the molecules). The dotted lines indicate the bases of the pyramids, and the dashed line is the position of the mirror plane (indicating that each isomer is a mirror image of the other.)

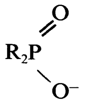
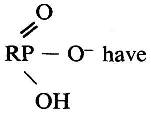
Tetrahedral structures

The great majority of phosphorus compounds, however, are not pyramidal but rather have four substituents arranged tetrahedrally about the phosphorus atom. These tetrahedral compounds include salts of the formally trivalent compounds—formed by utilizing the two *s* electrons and based on the phosphonium ion, PH_4^+ (in which H is hydrogen)—as well as the much larger group of truly pentavalent compounds. Pentavalent compounds can be formed because the phosphorus atom has available unused orbitals—called *d* orbitals—which, in the tetrahedral PX_4^+ state, readily accept electrons from other (donor) atoms to form a fifth bond to phosphorus. The donor atom may be in a separate molecule, in which case a pentacovalent species PX_5 results; several scores of compounds with this structure are known. But the utilization of *d* orbitals is most efficient when the donor atom is already bonded to phosphorus. The phosphorus-oxygen bond, for example, is stabilized in compounds such as the phosphine oxides, $R_3P{=}O$ (in which O is an oxygen atom, R represents an organic group, and the double line indicates a double bond), the *p* electrons of oxygen interacting with the vacant *d* orbitals of phosphorus to form a so-called pi bond. This type of multiple bonding stabilizes an adjacent negatively charged centre very effectively in the phosphinimines, $R_3P{=}NR$, and phosphinemethylenes, $R_3P{=}CR_2$ (C representing a carbon atom), as well as being responsible for the great stability of the phosphorus-oxygen double bond (P=O).

It is also instructive to compare these higher valence compounds of phosphorus to the comparable compounds of nitrogen. Nitrogen atoms, like phosphorus atoms, carry *s* electrons that they are able to use in forming salts—derivatives of the ammonium ion, NH_4^+, just as the related phosphorus compounds are derived from the phosphonium ion, PH_4^+. The nitrogen atom, however, does not have vacant *d* orbitals available for sharing donor electrons, so that pi-bonded compounds comparable to those formed by phosphorus do not occur. Thus, the amine oxides, compounds formulated as $R_3N^+{-}O^-$, cannot form a carbon–nitrogen double bond and are markedly less stable than the comparable phosphorus–oxygen compounds, which do form a phosphorus–oxygen double bond.

This exceptional strength of the phosphorus–oxygen double bond in the phosphoryl compounds has an important influence on much of phosphorus chemistry.

Importance of the phosphorus–oxygen double bond

Chemical reactivity. Trivalent phosphorus compounds are readily converted to their oxides and undergo many other reactions that generate products containing the phosphorus–oxygen double bond. Furthermore, the stabilizing effect of pi bonding can be exerted on more than one phosphorus–oxygen bond simultaneously. As a result, anions (negatively charged ions) such as that formulated  $R_2P(=O)O^-$ are unusually stable (in having two bonds, both of which can show pi bonding). Indeed, compounds formulated as $R_2P(O)OH$ are strong acids; *i.e.,* they readily give up a hydrogen ion to achieve the stable anionic condition. Even such anions as  $RP(=O)(OH){-}O^-$ have acidic properties and in neutral solution are largely dissociated to dianions, RPO_3^{2-}. In these anions, the various oxygen atoms are equivalent and share in the negative charge, while the various phosphorus–oxygen bonds are also equivalent and have equal amounts of partial double bond character.

Chemically, the phosphorus atom of an organophosphorus compound is a centre of high reactivity. Trivalent phosphorus has a lone pair of nonbonded *s* electrons, which gives it basic (the chemical opposite of acidic) character, as well as nucleophilic character (that is, a tendency to be attracted to positively charged groups). Nucleophilic substitution reactions proceed readily at both trivalent and pentavalent phosphorus centres. The only important exceptions to this generalization are the tertiary phosphine oxides, $R_3P{=}O$. These have no readily displaced substituent and, therefore, do not undergo substitution reactions of any kind. Furthermore, the doubly bonded phosphorus–oxygen group, $P{=}O$, does not undergo the ready addition reactions characteristic of the carbon–oxygen doubly bonded group ($C{=}O$, or carbonyl group), the behaviour of which in aldehydes and ketones is an important aspect of general organic chemistry. Thus, the tertiary phosphine oxides, $R_3P{=}O$, show little or none of the chemical reactivity associated with the ketones, $R_2C{=}O$.

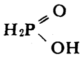
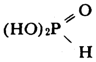
Analysis of phosphorus in compounds

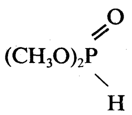
The phosphorus content of an organic compound is usually assayed (measured) by first subjecting the compound to reactions (hydrolysis and oxidation) that transform the phosphorus atoms to inorganic phosphate. Some compounds, especially those with phosphorus–carbon bonds, are degraded to inorganic phosphate only by powerful oxidizing agents; *e.g.*, concentrated perchloric acid at the boiling point. The inorganic phosphate formed is converted to a substance called phosphomolybdate, which can be determined either gravimetrically (by weight) or colorimetrically (by measurement of colour intensity).

MAJOR CLASSES

Organic derivatives of phosphorus acids. The organic derivatives of the acids based on phosphorus are derived from the six parent acids shown in Table 28. Conversion to organic derivatives can be considered formally to consist of replacement of any or all of the hydrogen atoms by organic groups (R). In addition, one or more of the oxygen atoms may be replaced by sulfur atoms, giving rise to the large family of thioacids of phosphorus.

Table 28: Trivalent and Pentavalent Forms of the Phosphorus Acids

trivalent form		pentavalent form	
name	formula	name	formula
Phosphinous acid	H_2POH	phosphine oxide	$H_3P{=}O$
Phosphonous acid	$HP(OH)_2$	phosphinic acid	 $H_2P(=O)OH$
Phosphorous acid	$(HO)_3P$	phosphonic acid	 $(HO)_2P(=O)H$
		phosphoric acid	$(HO)_3P{=}O$

The free acids themselves exist overwhelmingly in the pentavalent ($P = O$) forms (shown in the second column of the table), but esters and other derivatives of the trivalent are stable. If this generalization is not appreciated, the nomenclature can be very confusing. Trimethyl phosphite $(CH_3O)_3P$, for example, is a stable liquid that can readily be hydrolyzed (converted by water) to the dimethyl ester. This ester is commonly known, not unreasonably, as dimethyl phosphite, but it exists almost exclusively in the tetrahedral form,  $(CH_3O)_2P(=O)H$, and is, therefore, correctly termed dimethyl phosphonate.

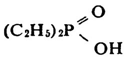
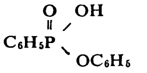
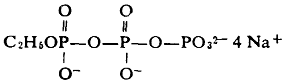
Most organic compounds of phosphorus formally are derivatives of one of the phosphorus acids—the exceptions being those compounds with three or more phosphorus–carbon or phosphorus–hydrogen bonds. The neutral organic compounds of phosphorus are described below, in sections devoted to the trivalent and pentavalent derivatives, respectively. This section deals specifically with the acid forms, including those acid esters in which the parent acid has more than one acidic group. (Most important by far of these are the esters of phosphoric acid.)

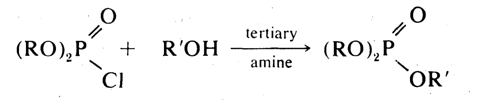
Organic phosphorus acids

Organic acids of phosphorus are moderately strong acids, with dissociation constants (the standard way of comparing acids) in the region of 10^{-1}–10^{-2}. Acids with two hydroxyl ($-OH$) groups have a second dissociation constant, usually between 10^{-7} and 10^{-8}. The lower molecular weight compounds are liquids that are soluble in water. The higher molecular weight acids and those containing aromatic groups (a particularly stable organic grouping) become progressively less soluble in cold water. Even these, however, generally dissolve in neutral or alkaline aqueous solution to form anions. As usual, solubility in organic solvents follows almost the reverse trend: it is higher for nonionized acids and for esters with high molecular weights than for those of low molecular weights. In Table 29 is collected a representative selection of acids and their derivatives based on simple organic groups; it illustrates the system of nomenclature now in use, as well as typical physical properties.

Phosphate esters—the esters of phosphoric acid—like other tetrahedral compounds of pentavalent phosphorus, are usually prepared by nucleophilic substitution of groups located on the phosphorus atom (that is, by attack of a negative group at the positive centre). Syntheses of simple derivatives are based on the readily available substance phosphorus oxychloride, $POCl_3$, which reacts consecutively to lose one, two, or three chloride ions. Triesters are formed when a dialkyl phosphorochloridate (prepared from two molecules of the alcohol to one of phosphorus oxychloride) reacts with an alcohol or phenol in the presence of a tertiary amine, as in the following chemical equation:

Table 29: Organic Phosphorus Acids and Derivatives

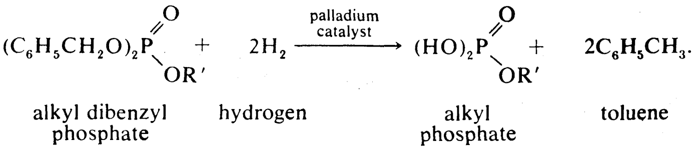
structure	name	melting point (°C)	boiling point (°C)
 $(C_2H_5)_2P(=O)OH$	diethylphosphinic acid	19	310
 $C_6H_5P(=O)(OH)OC_6H_5$	phenyl hydrogen phenylphosphonate	70–72	...
$(C_2H_5O)_3P{=}O$	triethyl phosphate	...	215
$C_2H_5OPOCl_2$	ethyl phosphorodichloridate	...	167
$C_6H_5OPO(OH)_2$	phenyl dihydrogen phosphate	97–98	...
$C_2H_5OPO_3^{2-}\ Mg^{2+}$	magnesium ethyl phosphate	...	...
$H_3\overset{+}{N}PO_3^{2-}\ Na^+$	sodium phosphoramidate	...	...
$(C_2H_5O)_2P(O)OPO(OC_2H_5)_2$	tetraethyl pyrophosphate	...	137/1mm
 $C_2H_5OP(=O)(O^-){-}O{-}P(=O)(O^-){-}O{-}PO_3^{2-}\ 4\ Na^+$	tetrasodium ethyl tripolyphosphate	...	...



$$(RO)_2P(=O)Cl + R'OH \xrightarrow[\text{amine}]{\text{tertiary}} (RO)_2P(=O)OR'$$

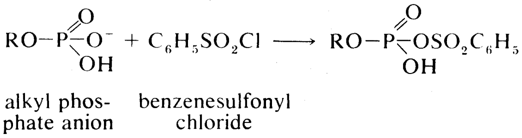
dialkyl phosphorochloridate alcohol trialkyl phosphate

This method provides an unambiguous route to unsymmetrical triesters and is readily adapted to the synthesis of monoesters if the first two alkyl groups (R in the above example) can be removed under conditions that leave the third group unaffected. The benzyl group ($C_6H_5CH_2-$) is a group of this kind frequently used in the preparation of phosphate esters. It can readily be removed by treating the benzyl ester with hydrogen, as in the following example:

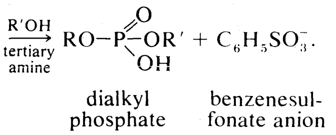


$$(C_6H_5CH_2O)_2\overset{O}{\underset{OR'}{P}} + 2H_2 \xrightarrow{\text{palladium catalyst}} (HO)_2\overset{O}{\underset{OR'}{P}} + 2C_6H_5CH_3.$$

Important variations of this type of reaction can be carried out with condensing agents, such as acid chlorides or anhydrides, on phosphate anions. These condensing agents convert the negatively charged oxygen atom to a group that is easily displaced. An example of such an easily displaceable group is the benzenesulfonate group (introduced by reaction with benzenesulfonyl chloride). The reactions involved are the following:

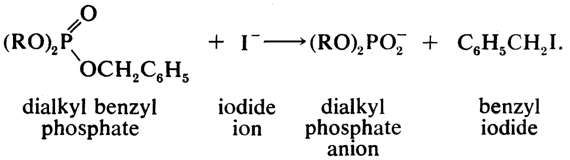


$$RO{-}\overset{O}{\underset{OH}{P}}{-}O^- + C_6H_5SO_2Cl \longrightarrow RO{-}\overset{O}{\underset{OH}{P}}{-}OSO_2C_6H_5$$



$$\xrightarrow[\text{tertiary amine}]{R'OH} RO{-}\overset{O}{\underset{OH}{P}}{-}OR' + C_6H_5SO_3^-.$$

Alternatively, a carbon–oxygen bond may be broken when a powerful but weakly basic nucleophile attacks an ester with a reactive alkyl group, as shown below:



$$(RO)_2\overset{O}{\underset{OCH_2C_6H_5}{P}} + I^- \longrightarrow (RO)_2PO_2^- + C_6H_5CH_2I.$$

Beginning in the 1950s the requirement for phosphate esters of known structure in the nucleotide field (see below) greatly stimulated work on methods of phosphorylation.

In their reactions, phosphate esters, especially the triesters, resemble the better known carboxylic esters. Strongly basic nucleophiles, such as hydroxide ions, readily effect nucleophilic displacement of groups on phosphorus atoms in diesters and triesters, as in this example:

$$(C_6H_5O)_3P{=}O \xrightarrow[\text{hydroxide ion}]{OH^-} (C_6H_5O)_2PO_2^- \xrightarrow{OH^-} C_6H_5OPO_3^{2-}.$$

triphenyl phosphate — diphenyl phosphate anion — phenyl phosphate dianion

Apart from their biochemical importance, organic phosphates play a definite role in the synthesis of organic chemicals containing phosphorus. Organic phosphates and related compounds also find important uses as corrosion and oxidation inhibitors; as plasticizers that confer fire resistance to textiles; as agents for the selective extraction of metals; and as insecticides. Some phosphorylating agents react rapidly and specifically to inhibit certain enzymes, particularly cholinesterase, which is essential for the transmission of nerve impulses. As a result, such compounds are extremely toxic to warm-blooded animals; known as nerve gases, they are particularly hazardous because they can be absorbed through the skin as well as by swallowing or breathing.

Biologically important phosphate esters

Phosphate esters play a vital part in the chemistry of the life process. The genetic code is carried by the deoxyribonucleic acids (DNA), which are high-molecular-weight phosphate esters, built up from four different subunits, called nucleotides. Nucleotides are the phosphate esters of nucleosides (comprised of sugars and organic bases), and in the nucleic-acid structure each nucleotide subunit is joined to the phosphate group of the next by a second ester linkage. The nucleotides themselves also take part in a large number of essential biochemical processes, not only in their monophosphate form but also as diphosphates and triphosphates. Nucleoside triphosphates, especially adenosine triphosphate (ATP), play an important role in the process of phosphate transfer (phosphorylation). This key reaction provides the chemical means by which an organism harnesses the chemical energy available from the degradation of food to do useful work. Phosphorylation is also an important biosynthetic reaction, because a large proportion of the compounds involved in metabolism are present in the organism as their phosphate esters; *e.g.*, the simple-sugar phosphates and phospholipids.

Several other types of phosphorus compounds are found in nature to a minor extent, including several with phosphorus–nitrogen bonds and a few with phosphorus–carbon bonds (phosphonates).

The acid esters have their own characteristic properties. The monesters, $ROPO(OH)_2$, for example, are stable to alkaline hydrolysis because they are converted to the unreactive dianions, $ROPO_3^{2-}$. At the same time, however, these monoesters are rapidly hydrolyzed under slightly acidic conditions in a unique reaction of the monoanion, which is the predominant ionic form under these conditions. Phosphate diesters are relatively unreactive. They show no special reactivity at any acidity and are hydrolyzed only slowly in alkali—under which conditions, as anions, their reaction with hydroxide ion is inhibited but not completely prevented by electrostatic repulsion; that is, repulsion between electrical charges of the same sign.

Certain structural characteristics confer high reactivity on phosphate esters. Diesters and triesters with their phosphorus atoms contained in a five-membered ring structure, for example, are hydrolyzed some 10^7 times as rapidly as the corresponding open-chain compounds. Also activated toward hydrolysis are many esters with neighbouring ionizable groups close enough to interact with the phosphorus centre. The best known example of this situation occurs in the ribonucleic acids, one of the two principal classes of nucleic acids. Ribonucleic acids are rapidly degraded by alkali under conditions in which the deoxyribonucleic acids, the other principal class, are quite stable. The only structural difference between the two classes of compounds is the presence of an extra hydroxyl group adjacent to the phosphate–ester linkage in the ribonucleic acids.

Trivalent organic phosphorus compounds. Compounds with three substituents bonded to a phosphorus atom have structures resembling those of the corresponding and more familiar nitrogen compounds, but the former group includes a much larger number of stable compounds with bonds to electronegative elements, such as oxygen and the halogens (fluorine, chlorine, bromine, and iodine). Nomenclature of the phosphorus compounds is more complicated than that of the nitrogen derivatives, which are named generally as amines. Naming of phosphorus compounds depends on whether phosphorus is joined solely to carbon or hydrogen atoms or both or whether it is bonded to an electronegative element. Compounds with bonds from phosphorus only to carbon or hydrogen are named as derivatives of phosphine, PH_3, by a system similar to that used for the amines. Thus, methylphosphine, CH_3PH_2, dimethylphosphine, $(CH_3)_2PH$, and trimethylphosphine, $(CH_3)_3P$, are primary, secondary, and tertiary phosphines, respectively. The metal salts are called phosphides, and the protonated forms (compounds to which a hydrogen ion is added) and alkylated forms (compounds to which an organic group is added) are called phosphonium compounds. Compounds with one or more bonds from phosphorus to any element other than carbon or hydrogen are named as derivatives of phosphinous, phosphonous, and phosphorous acids (see Table 28). Compounds with groups formulated as P—OR and P—SR are regarded as esters of the appropriate acid and thio (sulfur-containing) acid, respectively. Amino and halogen compounds are considered to be amides and acyl halides of the respective acids. A representative selection of trivalent phosphorus compounds based on simple organic groups is listed in Table 30. In all, a wide range of compounds is possible, with phosphorus making three bonds to any combination of the following groups: H, R (any organic group), NH_2, NHR, NR_2, OR, SR, SeR (Se being selenium), F, Cl, Br, and I. Compounds belonging to nearly all these

Nomenclature of trivalent phosphorus compounds

Table 30: Trivalent Phosphorus Compounds

structure	name	boiling point (°C)
$(C_2H_5)_4P^+ Br^-$	tetraethylphosphonium bromide	*
$(C_2H_5)_3P^+H Cl^-$	triethylphosphonium chloride	
$(C_2H_5)_3P$	triethylphosphine	127
$(C_2H_5)_2P—P(C_2H_5)_2$	tetraethyldiphosphine	
$(C_2H_5)_2PH$	diethylphosphine	85
$C_2H_5PH_2$	ethylphosphine	25
$(C_2H_5)_2P^- K^+$	potassium diethylphosphide	
$(C_2H_5)_2PCl$	diethylphosphinous chloride	60–70/15mm
$(C_2H_5)_2PN(C_2H_5)_2$	tetraethylphosphinous amide	181
$C_2H_5P(NH_2)_2$	ethylphosphonous diamide	—
$(C_2H_5)_2POC_2H_5$	ethyl diethylphosphinite	80–85/15mm
$C_2H_5P(OC_2H_5)_2$	diethyl ethylphosphonite	137–9
$(C_2H_5O)_3P$	triethyl phosphite	156–7
$(C_2H_5S)_3P$	triethyl phosphorotrithioite	140–3/18mm

*Decomposes at 300° C.

categories can be made by nucleophilic substitution reactions at the phosphorus atom, using the readily available phosphorus trihalides, especially phosphorus trichloride, PCl_3. Amines—and alcohols and phenols in the presence of added base—readily displace one or more chloride ions in a phosphorus chloride. Alkyl groups can be introduced by using organic magnesium compounds. A reaction sequence with examples of both reactions is:

$$C_2H_5MgCl \quad + \quad PCl_3 \longrightarrow C_2H_5PCl_2 \; + \; MgCl_2$$

ethylmagnesium chloride	phosphorus trichloride	ethylphosphonous dichloride	magnesium chloride

$$C_2H_5PCl_2 + 2C_2H_5OH \xrightarrow{\text{base}} C_2H_5P(OC_2H_5)_2 + 2HCl.$$

ethylphosphonous dichloride	ethyl alcohol	diethyl ethylphosphonite	hydrogen chloride

The lower phosphines, with phosphorus–hydrogen bonds, can be converted to metal salts, and these are readily alkylated by alkyl halides, as in the following example:

$$K^+PH_2^- \; + \; C_2H_5Br \; \longrightarrow \; C_2H_5PH_2 \; + \; KBr.$$

potassium phosphide	ethyl bromide	ethylphosphine	potassium bromide

These metal salts are easily accessible and useful as synthetic intermediates.

In many important reactions of trivalent phosphorus compounds, the lone pair of *s* electrons on phosphorus is attacked by electrophilic reagents. The most reactive such compounds are the phosphines themselves; electronegative substituents on phosphorus and increasing molecular weight both reduce reactivity. Many aliphatic and some aromatic phosphines, for example, are so readily oxidized by atmospheric oxygen that they are spontaneously flammable in air; these same compounds react explosively with reactive alkylating agents, such as methyl iodide; but triphenyl phosphite, $(C_6H_5O)_3P$, and triphenylphosphine, $(C_6H_5)_3P$, are oxidized and alkylated slowly and can be handled without special precautions. Similarly, triphenylphosphine is only weakly basic, although aliphatic tertiary phosphines (but not primary or secondary) are bases as strong as the corresponding amines.

Three reactions typical of the phosphines are illustrated by the following equations:

$$(C_2H_5)_3P + C_2H_5I \xrightarrow{\text{cold ether solvent}} (C_2H_5)_4P^+I^-,$$

triethylphosphine	ethyl iodide	tetraethylphosphonium iodide

$$(C_6H_{13})_3P \; + \; \tfrac{1}{2}O_2 \xrightarrow{\text{stand in air}} (C_6H_{13})_3P{=}O,$$

tri-*n*-hexylphosphine	oxygen	tri-*n*-hexylphosphine oxide

$$(C_2H_5)_3P \; + \; Br_2 \xrightarrow[\text{solvent}]{\text{cold benzene}} (C_2H_5)_3P^-Br^+ \; Br^-.$$

triethylphosphine	bromine	bromotriethylphosphonium bromide

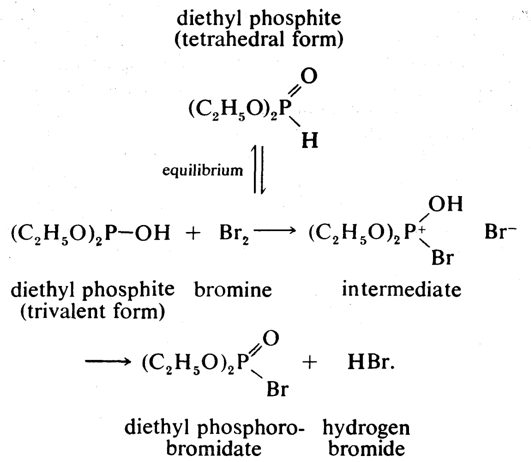
Many trivalent phosphorus compounds remove atoms of oxygen from various molecules, especially when they are attached to nitrogen or sulfur or to another phosphorus centre; for example,

$$(C_2H_5O)_3P \; + \; R_3\overset{+}{N}{-}O^- \longrightarrow (C_2H_5O)_3P{=}O \; + \; R_3N.$$

triethyl phosphite	tertiary amine oxide	triethyl phosphate	tertiary amine

Many classes of trivalent compounds also form series of complex coordination compounds with metal ions.

As stated above, hydroxyl compounds of trivalent phosphorus exist almost exclusively in the tetrahedral P═O form. Extremely small amounts of the trivalent hydroxyl forms are present in equilibrium, however (often much less than one part per million), and these are thought to be responsible for the observed reactions of the compounds with reactive electrophilic reagents. An example is given below:



Phosphonium salts

The availability for bonding of the *d* orbitals of phosphorus has a decisive effect on the chemistry of phosphonium compounds. Generally, the reactions of phosphonium compounds are more important than those of ammonium compounds, and their chemical behaviour is quite different. Some of these compounds form a fifth bond to phosphorus, as in the following example:

$$(C_6H_5)_4P^+ \; Br^- \; + \; C_6H_5Li \; \longrightarrow \; (C_6H_5)_5P \; + \; LiBr.$$

tetraphenylphosphonium bromide	phenyllithium	pentaphenylphosphorane	lithium bromide

Usually, the fifth bond is a pi bond, as in the following:

$$(C_6H_5)_3\overset{+}{P}CH_2C_6H_5Br^- + C_6H_5Li$$

benzyltriphenylphosphonium bromide	phenyllithium

$$\longrightarrow \; (C_6H_5)_3P{=}CHC_6H_5 \; + \; LiBr \; + \; C_6H_6.$$

triphenylbenzylidenephosphorane	lithium bromide	benzene

Most commonly the multiple bond formed is the very stable P═O bond. Alkoxyphosphonium compounds, for example, are rapidly dealkylated by most nucleophiles to give phosphine oxides:

$$(C_2H_5)_3\overset{+}{P}OC_2H_5 \; Br^- \longrightarrow (C_2H_5)_3P{=}O + C_2H_5Br.$$

triethylethoxyphosphonium bromide	triethylphosphine oxide	ethyl bromide

As a result of this reaction, phosphoryl compounds are the final products when any aliphatic esters of trivalent phosphorus acids react with normal alkylating agents, as shown in the following example:

$$(C_2H_5)_2POC_2H_5 + C_2H_5Br \xrightarrow{\text{2 steps}} (C_2H_5)_3P{=}O + C_2H_5Br.$$

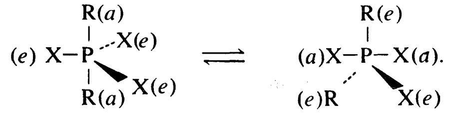
ethyl diethylphosphinite   ethyl bromide   triethylphosphine oxide   ethyl bromide

When the same alkyl group is involved in both ester and alkylating agent, as in the example shown, the alkylating agent is regenerated at the end of the reaction, and only catalytic amounts of it are required. In many instances, the same overall reaction can be brought about simply by strong heating, without the addition of any alkylating agent at all.

Phosphonium compounds are hydrolyzed by strong alkali in an unusual reaction in which an organic group, *R*, is displaced from phosphorus, apparently as the anion R^-. This remarkable result is feasible only because of the great stability of the P═O group of the product and because a mechanism is possible in which two molecules of hydroxide cooperate to break the carbon–phosphorus bond.

Pentavalent organic phosphorus compounds. The organic compounds of pentavalent phosphorus are of two structural types. A small number have five single bonds to phosphorus and trigonal bipyramidal geometry (that is, a structure formed by two pyramids joined base to base). Examples are pentaphenylphosphorane, $(C_6H_5)_5P$, and pentaethoxyphosphorane, $(C_2H_5O)_5P$. In compounds of this structure, two substituents are in axial (*a*) positions, and three are in equatorial positions (*e*), as shown in the structures below. Compounds with substituents in one position can in some cases be distinguished from those with substituents in the other. Complications arise, however, because axial and equatorial substituents can exchange positions by a process of molecular reorganization known as pseudorotation. An example of pseudorotation is shown below:

Two structural types

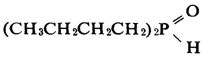
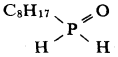
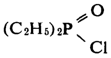
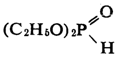


In this example, the two axial substituents (R) become equatorial substituents by changing of the bond positions. In these diagrams, ordinary lines represent bonds in the plane of the paper, dotted lines are bonds extending to the rear, and the wedges indicate bonds extending forward. Compounds with the PX_5 structure are generally very reactive. They readily revert to tetrahedral compounds, usually by way of an initial reversible ionization of the type $PX_5 \rightleftharpoons PX_4^+ + X^-$. An equilibrium of this sort is possible for any phosphonium system, and whether a particular compound exists predominantly as PX_5 or as $PX_4^+X^-$ depends in borderline cases on the environmental conditions (*e.g.*, solvent, temperature, and so on), as well as on the structure of the molecule. Compounds of the PX_5 structure have electrophilic (Lewis-acid) character also, and they may accept electrons from suitable donors to form the rare six-coordinate compounds PX_6^-. One or two organic examples of this class are known.

Compounds with pi bonds

The great majority of pentavalent organophosphorus compounds and, thus, also of all organic phosphorus compounds have four single bonds to phosphorus, one or more of which is reinforced by an additional bond, a pi bond. For purposes of nomenclature these compounds are classified in the same way as the corresponding trivalent derivatives: those with three bonds to carbon or hydrogen have names based on the parent phosphines, and the remainder are named as derivatives of phosphorus acids—in this case phosphinic or phosphonic acids. The range of structures known is essentially that given above for trivalent compounds, with the addition of the fourth bond, which is commonly a phosphorus to oxygen bond, P═O, or phosphorus–sulfur bond, P═S, but may also be a phosphorus–nitrogen or phosphorus–carbon bond, P═NR or P═CR₂. The selection of compounds listed in Table 31 illustrates the physical properties of these compounds, as well as the ways of naming them.

Table 31: Pentavalent Phosphorus Compounds

structure	name	melting point (°C)	boiling point (°C)
$(C_6H_5)_5P$	pentaphenylphosphorane	124	—
$(C_2H_5)_3P{=}O$	triethylphosphine oxide	50	238–40
$(C_2H_5)_3P{=}S$	triethylphosphine sulfide	94	—
$(C_2H_5)_3P{=}NC_2H_5$	tetraethylphosphine imine (or imide)	—	—
$(C_2H_5)_3P{=}C(C_2H_5)_2$	pentaethylmethylenephosphorane	—	—
	dibutylphosphinous acid (or dibutylphosphine oxide)	58–60	—
	octylphosphine oxide	46–48	—
	diethylphosphinic chloride	—	109/16mm
$C_2H_5P(O)Cl_2$	ethylphosphonic dichloride	—	34/3mm
$(C_2H_5)_2P(O)N(C_2H_5)_2$	tetraethylphosphinic amide	—	134/16mm
$(C_2H_5)_2P(O)OC_2H_5$	ethyl diethylphosphinate	—	92/14mm
	diethyl phosphonate	—	187–88
$C_2H_5PO(OH)_2$	ethylphosphonic acid	61–62	—
$C_2H_5PO(OC_2H_5)_2$	diethyl ethylphosphonate	—	198
$(C_2H_5O)_3P{=}O$	triethyl phosphate	—	215–16
$(C_2H_5O)_3P{=}S$	*O,O′,O″*-triethyl phosphorothioate	—	105/20mm
$(C_2H_5S)_3P{=}S$	triethyl phosphorotetrathioate	—	110/0.2mm

Tetrahedral compounds of pentavalent phosphorus are generally prepared by reactions involving nucleophilic substitution at phosphorus or oxidation of the appropriate trivalent derivatives. Many reactions of trivalent compounds that give phosphonium compounds as the initial product also lead eventually to phosphoryl derivatives. Several such reactions have been described above. One that is of considerable importance in organic synthesis is the Wittig reaction (named for its discoverer, the German chemist Georg Wittig), in which the oxygen atom of a carbonyl group (carbon and oxygen atoms forming a double bond, C═O) of a ketone or aldehyde is replaced by a doubly bonded carbon atom (methylene group) from a methylenephosphorane, producing an olefin, usually in good yield. The Wittig reagent (the methylenephosphorane) is generated from a phosphonium compound by removing a proton with a strongly basic reagent:

The Wittig reaction

$$(C_6H_5)_3\overset{+}{P}CH_2C_6H_5Cl^- + CH_3(CH_2)_3Li \longrightarrow$$

triphenylbenzylphosphonium chloride   *n*-butyllithium

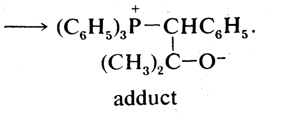
$$(C_6H_5)_3P = CHC_6H_5\ \{\text{or } (C_6H_5)_3\overset{+}{P} - \overset{-}{C}H_6CH_5\}.$$

triphenyl- benzylidenephosphorane

Although the negative charge in the so-called ylide form is largely neutralized by the adjacent positive phosphorus centre (P^+), the unsaturated carbon atom involved remains strongly nucleophilic. In particular, it adds readily to the carbonyl group of aldehydes and ketones, as shown in the example below:

$$(C_6H_5)_3P{=}CHC_6H_5 + (CH_3)_2C{=}O$$

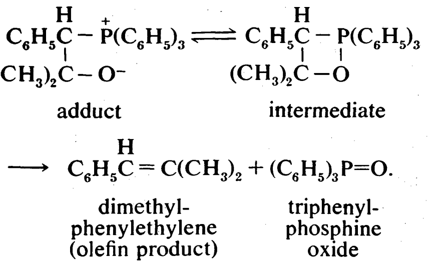
triphenylbenzylidenephosphorane   acetone



$$\longrightarrow (C_6H_5)_3\overset{+}{P}-\underset{(CH_3)_2C-O^-}{\underset{|}{C}}HC_6H_5.$$

adduct

The initial adduct rapidly decomposes to the final products, which are the desired olefin and a phosphine oxide, perhaps by way of a cyclic intermediate, as shown below:



$$\begin{array}{ccc} C_6H_5\overset{H}{C}-\overset{+}{P}(C_6H_5)_3 & \rightleftharpoons & C_6H_5\overset{H}{C}-P(C_6H_5)_3 \\ | & & |\qquad| \\ (CH_3)_2C-O^- & & (CH_3)_2C-O \end{array}$$

adduct intermediate

$$\longrightarrow C_6H_5\overset{H}{C}=C(CH_3)_2+(C_6H_5)_3P{=}O.$$

dimethylphenylethylene (olefin product) triphenylphosphine oxide

As in earlier examples, the formation of the stable P═O group is an important driving force for this reaction.

The important reactions undergone by the tetrahedral pentavalent organophosphorus compounds have already been described for the phosphate esters and their derivatives. The phosphorus–carbon bond is a strong one, and, with the exception of the Wittig reaction, electronegative substituents on phosphorus are the ones normally involved in chemical reactions. (A.J.K.)

Organic sulfur compounds

Organic sulfur compounds constitute a diverse and important subdivision of the class of organic substances. They are widely distributed in nature, often betraying their presence by the strong odours they impart (*e.g.*, to crude petroleum, to certain plants, and to animal secretions). Sulfur-containing amino acids—cysteine, cystine, methionine, and taurine—are important components of biologically important substances, including hormones, enzymes, and coenzymes. Synthetic organic sulfur compounds include numerous insecticides, pharmaceuticals, dyes, solvents, and agents used in flotation processes for refining ores, in improving the performance of lubricating oils, in preparing rubbers, and in making rayon.

After carbon and hydrogen, the elements most commonly occurring in organic compounds are oxygen and nitrogen; sulfur is neither extremely common nor extremely rare, but its effects are sufficiently distinctive that organic sulfur compounds comprise a recognized area of specialized study.

Basic characteristics

The sulfur atom. An uncombined sulfur atom has 16 electrons, 10 of which completely fill the 1*s*, 2*s*, and 2*p* orbitals; of the remaining six (*i.e.*, the valence electrons), two occupy the 3*s* orbital and four are present in 3*p* orbitals: The distribution of electrons in a sulfur atom is abbreviated to $1s^22s^22p^63s^23p^4$. An element closely resembling sulfur in electron configuration is oxygen, which has eight electrons in the configuration $1s^22s^22p^4$. The identical distribution, s^2p^4, of the valence electrons in the two elements leads to many similarities in their chemical behaviour.

Table 32: Organic Sulfur Compounds and Corresponding Organic Oxygen Compounds

sulfur compounds			oxygen compounds		
group of compounds	characteristic structural unit	example	group of compounds	characteristic structural unit	example
Thiols (aliphatic)	—SH	methanethiol	alcohols	—OH	methanol
Thiols (aromatic)	—SH	benzenethiol (thiophenol)	phenols	—OH	phenol
Sulfides	—S—	dimethyl sulfide	ethers	—O—	dimethyl ether
Disulfides	—S—S—	dimethyl disulfide	peroxides	—O—O—	dimethyl peroxide
Thioaldehydes	$-\overset{\overset{S}{\|}}{C}-H$	ethanethial (thioacetaldehyde)	aldehydes	$-\overset{\overset{O}{\|}}{C}-H$	ethanal (acetaldehyde)
Thioketones	$-\overset{\overset{S}{\|}}{C}-$	dimethyl thione (thioacetone)	ketones	$-\overset{\overset{O}{\|}}{C}-$	2-propanone (acetone)
Thiolocarboxylic acids	$-\overset{\overset{O}{\|}}{C}-SH$	thioloacetic acid			
Thionocarboxylic acids	$-\overset{\overset{S}{\|}}{C}-OH$	thionoacetic acid	carboxylic acids	$-\overset{\overset{O}{\|}}{C}-OH$	acetic acid
Dithiocarboxylic acids	$-\overset{\overset{S}{\|}}{C}-SH$	dithioacetic acid			

Important differences, however, arise from the fact that the valence electrons of sulfur interact less strongly with the nucleus than do those of oxygen and that the third electron shell of sulfur includes five *d* orbitals. Although these *d* orbitals are not occupied in the uncombined sulfur atom, their availability makes possible the formation of types of compounds not formed by oxygen. In these compounds, the 3*s* and 3*p* orbitals can, in effect, blend together into a new set of four "hybrid" orbitals that can interact with three or four other atoms, or the 3*s*, the 3*p*, and two of the 3*d* orbitals can hydridize to form a set of orbitals that participate in the formation of hexacovalent compounds.

In Table 32, members of several important groups of organic sulfur compounds are compared with the oxygen compounds possessing similar structures. Table 33 lists some of the most important types of sulfur compounds that have no counterparts among oxygen compounds.

Table 33: Organic Sulfur Compounds with No Oxygen Analogues

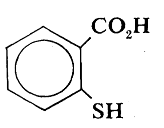
group of compounds	characteristic structural unit	example
Trisulfides	—S—S—S—	dimethyl trisulfide
Polysulfides	(—S—)*n*; *n* = 4,5,6, . . .	dimethyl tetrasulfide, dimethyl pentasulfide, etc.
Sulfoxides	$-\overset{\overset{O}{\|}}{S}-$	dimethyl sulfoxide
Sulfones	$-\underset{\underset{O}{\|}}{\overset{\overset{O}{\|}}{S}}-$	dimethyl sulfone
Sulfenic acids	—S—OH	methanesulfenic acid
Sulfinic acids	$-\overset{\overset{O}{\|}}{S}-OH$	methanesulfinic acid
Sulfonic acids	$-\underset{\underset{O}{\|}}{\overset{\overset{O}{\|}}{S}}-OH$	methanesulfonic acid

ORGANIC COMPOUNDS OF BIVALENT SULFUR

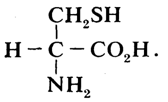
Thiols. The organic sulfur compounds most closely resembling their oxygen analogues are the thiols, also called mercaptans. The functional group of the thiols is the mercapto group (—SH), comparable to the hydroxyl group (—OH) present in alcohols and phenols. In preferred names for thiols, the suffix -thiol is appended to the name of the appropriate hydrocarbon. If another group is designated by a suffix, the prefix mercapto- is used. A few thiols are named by using the prefix thio- to denote the replacement of an oxygen atom of a related compound by a sulfur atom, and a few have names that do not convey structural information. Examples are as follows:

CH_3SH methanethiol

$HSCH_2CO_2H$ mercaptoacetic acid



thiosalicylic acid

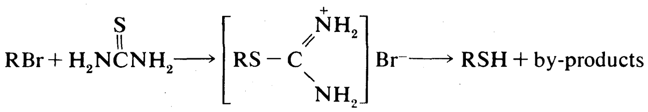


$$H-\underset{NH_2}{\overset{CH_2SH}{C}}-CO_2H.$$

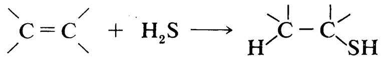
cysteine

Natural thiols

A few thiols are found in nature: crude petroleum contains methanethiol, ethanethiol, and other members of the group. Although they constitute a small fraction of most petroleums, thiols are a serious nuisance. They impart objectionable odours, they are corrosive to the equipment and interfere with the action of catalysts used in refining processes, and their combustion creates the noxious gas sulfur dioxide. Small amounts of thiols are converted into chemical products, but no uses require the amounts that could be recovered from crude oil.



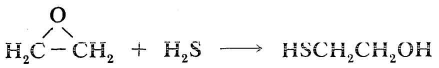
alkyl bromide · thiourea · alkylthiuronium bromide · alkanethiol



olefin or alkene · hydrogen sulfide · thiol

$$ROH + H_2S \xrightarrow{\text{catalyst}} RSH + H_2O$$

alcohol · hydrogen sulfide · thiol · water



ethylene oxide · hydrogen sulfide · 2-mercaptoethanol or 2-hydroxyethanethiol

$$CH_2=CHC\equiv N + H_2S \longrightarrow HSCH_2CH_2C\equiv N.$$

acrylonitrile · hydrogen sulfide · 3-mercapto-propionitrile

Methanethiol arises from the bacterial decomposition of proteins, such as albumin or gelatin, and butanethiol is present in the defensive secretion of the skunk. A thiol group plays an important role in the natural functions of several proteins and of coenzyme A, which participates in many metabolic reactions.

Preparation. Several methods are known for the preparation of thiols. Thiols in which the mercapto group is not attached to an aromatic ring may be made according to the reactions represented by the following equations:

$$RBr + NaSH \longrightarrow RSH + NaBr$$

alkyl bromide · sodium hydrosulfide · alkanethiol · sodium bromide

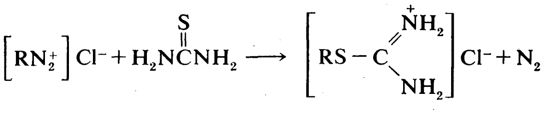
Other methods must be used to make aromatic thiols because of the different nature of reactions typical of aromatic compounds in general. Applicable procedures include those formulated as follows:

$$RSO_2Cl + 3Zn + 5HCl \longrightarrow RSH + 3ZnCl_2 + 2H_2O$$

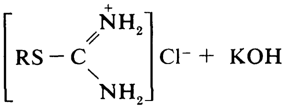
aromatic sulfonyl chloride · zinc · hydrochloric acid · aromatic thiol · zinc chloride · water

$$RSO_2Cl + 3SnCl_2 + 5HCl \longrightarrow RSH + 3SnCl_4 + 2H_2O$$

aromatic sulfonyl chloride · tin(II) chloride · hydrochloric acid · aromatic thiol · tin(IV) chloride · water



aromatic diazonium chloride · thiourea · arylthiuronium chloride · nitrogen



arylthiuronium chloride · potassium hydroxide

$$\longrightarrow RSH + H_2NC\equiv N + KCl + H_2O.$$

aromatic thiol · cyanamide · potassium chloride · water

Reactions. In most of their reactions, thiols resemble alcohols or phenols, the differences being those of degree rather than kind. Thiols react more completely with alkalies, forming salts, than do the corresponding hydroxyl compounds; that is, the thiols are stronger acids, although very weak in comparison to, say, acetic acid. Aromatic thiols are stronger acids than the aliphatic thiols, just as phenols are stronger acids than alcohols. In the presence of salts of heavy metals (as mercury, lead, zinc, or copper), thiols form mercaptides, soluble in organic solvents (as ether, isopropyl alcohol, benzene, chloroform) but insoluble in water. The characteristic formation of these compounds from mercury salts led to the name mercaptan (Latin *mercurium captans,* "seizing mercury"). The mercury mercaptide, sodium ethylmercurithiosalicylate (thimerosal, Merthiolate), is a well-known germicide.

Thiols form sulfides (thioethers) and thioesters in reactions similar to those of hydroxyl compounds. They react with aldehydes and ketones to yield thioacetals and thioketals, respectively; sulfur-containing compounds of this type are more readily formed and more stable than the oxygen compounds and are useful in suppressing the reactivity of carbonyl groups while chemical reactions are performed on another part of a molecule. Dithiols are useful for this purpose. Thiols combine with unsaturated compounds, particularly those that have a carbonyl group adjacent to the double bond, forming sulfides.

Thiols differ from hydroxyl compounds in their reactions with oxidizing agents (as oxygen, iodine, hydrogen peroxide). Alcohols usually are converted to aldehydes or ketones, but thiols are transformed into disulfides; further oxidation results in formation of sulfonic acids:

$$2RSH + H_2O_2 \longrightarrow RSSR + 2H_2O$$

thiol · hydrogen peroxide · disulfide · water

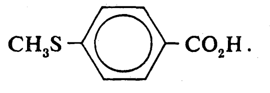
$$RSSR + 5H_2O_2 \longrightarrow 2RSO_3H + 4H_2O.$$

disulfide · hydrogen peroxide · sulfonic acid · water

Sulfides and sulfonium salts. Compounds in which two organic groups are bonded to a sulfur atom are called sulfides; the structurally related oxygen compounds are ethers. The organic groups may be both alkyl, both aryl, or one of each. If no other functional group is present in the molecule, sulfides are named as such: dimethyl sulfide is CH_3-S-CH_3, methyl phenyl sulfide is $CH_3-S-C_6H_5$, and diphenyl sulfide is $C_6H_5-S-C_6H_5$. If another functional group forms part of the molecule, a sulfide group is designated by the particle -thio-:

$HO_2CCH_2SCH_2CO_2H$

thiodiacetic acid



4-(methylthio)benzoic acid

Properties of sulfides

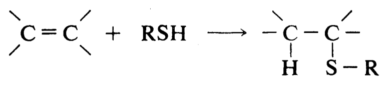
The sulfur atom in sulfides does not cause their physical properties to differ much from those of hydrocarbons of similar molecular size and shape. The sulfides have low solubilities in water, but they are miscible with many organic solvents; they are colourless liquids or solids, many of them possessing odours that are unpleasant, though not as intense as those of the thiols.

Certain sulfides occur in nature: garlic contains diallyl sulfide, and many proteins contain the amino acid methionine, which is a sulfide.

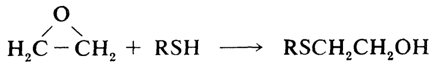
Preparation. Sulfides most often are prepared from thiols; just as preparing a thiol usually involves replacing one of the two hydrogen atoms of hydrogen sulfide, preparation of a sulfide is the replacement of the second, often by similar procedures. Examples are:

$$RSNa + R'Br \longrightarrow RSR' + NaBr$$

thiol sodium salt　alkyl bromide　sulfide　sodium bromide



olefin　thiol　sulfide



ethylene oxide　thiol　alkyl hydroxyethyl sulfide

$$CH_2=CHC\equiv N + RSH \longrightarrow RSCH_2CH_2C\equiv N.$$

acrylonitrile　thiol　3-alkylthiopropionitrile

Production of chlorine-containing sulfides

The reaction of certain olefins with disulfur dichloride produces chlorine-containing sulfides; the best known example of this process is the manufacture of mustard gas, a blister-forming chemical warfare agent, from ethylene. This reaction is not general, however; the products obtained from other olefins are disulfides or mixtures of monosulfides and disulfides.

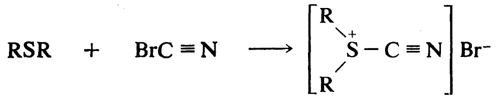
$$2C_2H_4 + S_2Cl_2 \longrightarrow S(CH_2CH_2Cl)_2 + S$$

ethylene　disulfur dichloride　mustard gas (bis[2-chloroethyl] sulfide)　sulfur

$$2CH_2CH=CH_2 + S_2Cl_2 \longrightarrow ClC_3H_6SSC_3H_6Cl$$

propylene　disulfur dichloride　bis(chloropropyl) disulfide

Reactions. Sulfides combine with chlorine, bromine, iodine, or salts of heavy metals to form crystalline compounds. Cyanogen bromide forms an unstable compound that decomposes by breaking a carbon–sulfur bond:

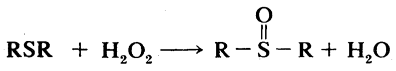


sulfide　cyanogen bromide　unstable sulfonium compound

$$\longrightarrow RBr + RSC\equiv N.$$

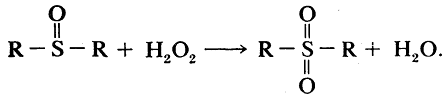
alkyl bromide　alkyl thiocyanate

Oxidizing agents convert sulfides into sulfoxides; if a sufficient amount of the oxidizing agent is present, the sulfoxides undergo oxidation to sulfones:



$$RSR + H_2O_2 \longrightarrow R-\overset{\overset{O}{\|}}{S}-R + H_2O$$

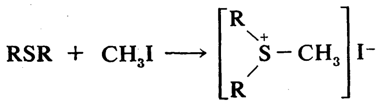
sulfide　hydrogen peroxide　sulfoxide　water



$$R-\overset{\overset{O}{\|}}{S}-R + H_2O_2 \longrightarrow R-\underset{\underset{O}{\|}}{\overset{\overset{O}{\|}}{S}}-R + H_2O.$$

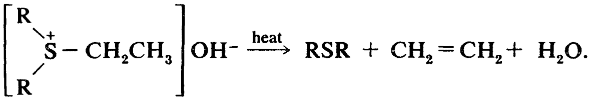
sulfoxide　hydrogen peroxide　sulfone　water

Compounds such as methyl iodide react with sulfides to produce sulfonium salts, in which three organic groups are attached to the positively charged sulfur atom. This process has many parallels in organic chemistry, especially in the reactions of amines and phosphines; it occurs less readily with ethers.



sulfide　methyl iodide　sulfonium iodide

The formation of sulfonium salts from sulfides and alkyl halides can be reversed by heat. The heat-induced decomposition of quaternary ammonium hydroxides (*i.e.*, compounds in which four organic groups are bonded to a positively charged nitrogen atom) is exactly duplicated in the case of sulfonium hydroxides:



sulfonium hydroxide　sulfide　ethylene　water

Disulfides and polysulfides. Disulfides have the structure $R-S-S-R'$, in which R and R′ represent organic groups of any kind. They are named by designating the groups attached to the sulfur atoms or by use of the particle -dithio-:

CH_3SSCH_3

dimethyl disulfide

$CH_3SSCH_2CO_2H$

methyldithioacetic acid

$HO_2CCH_2SSCH_2CO_2H.$

dithiodiacetic acid

Disulfides. The formation of disulfides by oxidation of thiols or by the reaction of disulfur dichloride and olefins has already been mentioned. A few disulfides have been isolated from natural sources, such as onion, garlic, and asafetida. The amino acid cystine, a disulfide, is an important component of many proteins: the sulfur–sulfur bond is partly responsible for maintaining the molecules in shapes essential for their biological activity. The coenzyme lipoic acid, a disulfide necessary for growth of certain microorganisms, may also be involved in photosynthesis and other vital processes of plants and animals.

Most important reactions

In the most important reactions of disulfides, the bond between the two sulfur atoms is broken, as by reducing agents, including hydrogen; by strong alkalies; or by alkyl halides:

$$RSSR + H_2 \longrightarrow 2RSH$$

disulfide　hydrogen　a thiol

$$RSSR + KOH \longrightarrow RSH + (RSO)K$$

disulfide　potassium hydroxide　a thiol　alkanesulfenate (salt)

$$\mathrm{RSSR + 4RI \xrightarrow{HgI_2} 2(R_3S^+)I^- + I_2}\,.$$

disulfide	alkyl iodide	sulfonium iodide	iodine

Chlorine or bromine splits the —S—S— bond of aromatic disulfides but attacks the adjacent carbon atom of an aliphatic disulfide:

$$\mathrm{C_6H_5SSC_6H_5 + Br_2 \longrightarrow 2C_6H_5SBr}$$

diphenyl disulfide	bromine	benzenesulfenyl bromide

$$\mathrm{CH_3SSCH_3 + Cl_2 \longrightarrow CH_3SSCH_2Cl + HCl}\,.$$

dimethyl disulfide	chlorine	chloromethyl methyl disulfide	hydrogen chloride

Polysulfides. The tendency to form chains in which atoms of the same element are joined together by covalent bonds, especially pronounced with carbon, is shown to some extent by sulfur. Organic compounds containing such chains of three, four, or more sulfur atoms are called polysulfides. They are less stable than the corresponding carbon compounds, and their properties are not as fully understood. A group of organic polysulfides has received considerable attention as rubbers and as solid propellants for rockets; these are the Thiokols, which are polymeric compounds, having molecules composed of long chains of repeating units built up from simpler compounds. An example is provided by the product made from ethylene dichloride and sodium tetrasulfide:

Thiokols

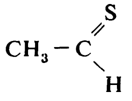
$$n\mathrm{ClCH_2CH_2Cl} + n\mathrm{Na_2S_4}$$

ethylene dichloride	sodium tetrasulfide

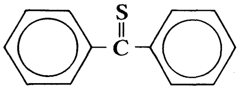
$$\longrightarrow \mathrm{(CH_2CH_2S_4)}_n + 2n\mathrm{NaCl}\,.$$

Thiokol A, a polymer	sodium chloride

Thiocarbonyl compounds. The thiocarbonyl functional group (>C=S), analogous to the carbonyl group, forms part of the structure of thioaldehydes and thioketones. These compounds are named by analogy with the corresponding oxygen compounds:

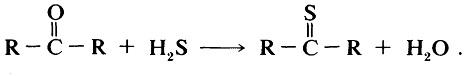


thioacetaldehyde or ethanethial



thiobenzophenone or diphenyl thione

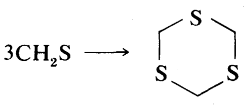
The thiocarbonyl compounds have exceptionally repulsive odours. The general method of preparation, from carbonyl compounds and hydrogen sulfide in the presence of acidic catalysts, is not difficult:



$$\mathrm{R-\overset{\overset{\displaystyle O}{\|}}{C}-R + H_2S \longrightarrow R-\overset{\overset{\displaystyle S}{\|}}{C}-R + H_2O}\,.$$

carbonyl compound	hydrogen sulfide	thiocarbonyl compound	water

The reactions of thiocarbonyl compounds are similar to those of their oxygenated relatives, although they show a greater tendency to undergo cyclotrimerization, in which three molecules unite to form a ring structure:



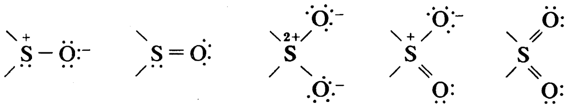
thioformaldehyde trithiane

Although the monomers are blue, the cyclic trimers (trithianes) are colourless; the latter somewhat resemble the dialkyl sulfides in their reactions.

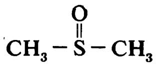
COMPOUNDS CONTAINING POLYVALENT SULFUR

Sulfur–oxygen double bonds

Sulfoxides and sulfones. Two groups of organic sulfur compounds that have no counterparts among organic oxygen compounds are the sulfoxides and the sulfones. In some respects, these compounds resemble carbonyl compounds, containing the group >C=O, in which the carbon and the oxygen atom each interact with eight electrons, four of which are shared in the double bond. Doubly bonded structures for the sulfinyl (>SO) and sulfonyl (>SO$_2$) groups present in sulfoxides and sulfones, respectively, require that the sulfur atom interact with 10 electrons in sulfoxides or with 12 in sulfones. Because these numbers are larger than eight, the number of electrons usually associated with each atom (except hydrogen) in stable covalent compounds, has been the basis of opposition to these formulations. Covalent bond formation by a sulfur atom, however, involves electrons in the third shell, which is larger than the second shell (the one engaged in carbon and oxygen bonding) and has additional orbitals. It is considered that these orbitals (the *d* orbitals) participate in the bonding in sulfoxides and sulfones and that the bonds differ from those of the carbonyl group, showing considerable polarity (separation of electrical charges). Their structures are regarded as intermediate between those with single bonds of high polarity and those with double bonds of low polarity but high electron density on the sulfur atom.

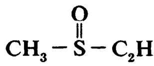


Sulfoxides are named by simply designating the two organic groups attached to the >SO group, with the word sulfoxide, or by forming a prefix from the name of the simpler of the two groups and the particle -sulfinyl-, as shown in the examples:



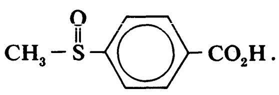
$$\mathrm{CH_3-\overset{\overset{\displaystyle O}{\|}}{S}-CH_3}$$

dimethyl sulfoxide



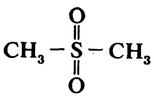
$$\mathrm{CH_3-\overset{\overset{\displaystyle O}{\|}}{S}-C_2H_5}$$

ethyl methyl sulfoxide



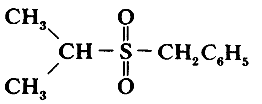
4-(methylsulfinyl) benzoic acid

The nomenclature of sulfones is like that of sulfoxides; the particle -sulfonyl- is used in complicated structures:

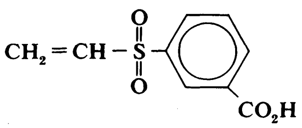


$$\mathrm{CH_3-\underset{\underset{\displaystyle O}{\|}}{\overset{\overset{\displaystyle O}{\|}}{S}}-CH_3}$$

dimethyl sulfone



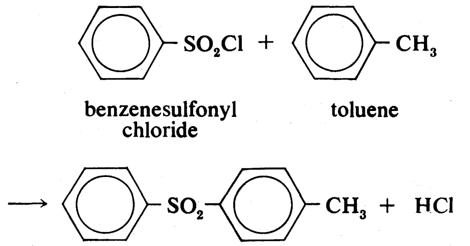
benzyl isopropyl sulfone



3-(vinylsulfonyl) benzoic acid

Most sulfoxides are colourless liquids or solids with low melting points; they are more soluble in water and possess much higher boiling points than hydrocarbons or carbonyl compounds of similar molecular size. Sulfoxides of low molecular weight have faint odours and tastes described as metallic or garlic-like. In general, sulfones are colourless, crystalline solids at room temperature.

Occurrence and preparation. Several sulfoxides are found in the turnip, garlic, and several species of mustard; another has been isolated from the cockroach. Oxidation of sulfides to sulfoxides and sulfones by hydrogen peroxide has been mentioned; certain aromatic sulfones can be made by the reaction of sulfonyl chlorides (see below *Sulfonic acids*) with aromatic hydrocarbons.



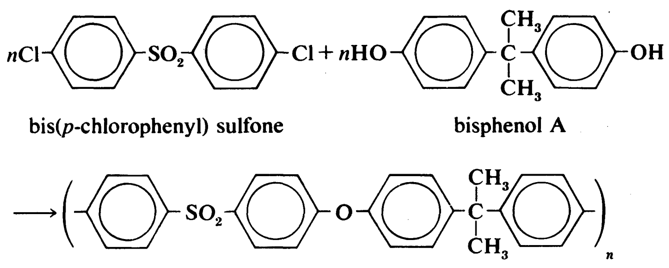
benzenesulfonyl chloride — toluene

phenyl *p*-tolyl sulfone — hydrogen chloride

The reaction of a metal salt of a sulfinic with a halogen compound is sometimes used to make sulfones.

Polymeric sulfones

A class of polymeric sulfones results from the reaction of sulfur dioxide with olefins. These products have not been found useful, although members of a different class of polysulfones have properties that make them valuable as wire coatings. The latter group is made from chlorine-containing aromatic sulfones and a compound (bisphenol A) of two molecules of phenol linked through a hydrocarbon group:

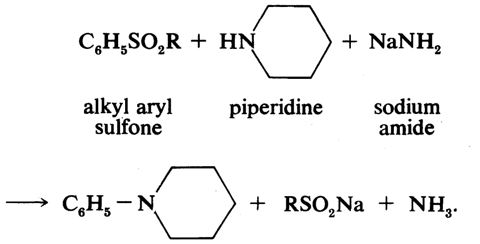


bis(*p*-chlorophenyl) sulfone — bisphenol A

polymer

Reactions. Sulfoxides can be converted to sulfides by powerful reducing agents, such as lithium aluminum hydride, hydriodic acid, or zinc in the presence of sulfuric acid. Sulfoxides are very weak bases, forming salts with strong acids (*e.g.*, hydrochloric acid); they are also very weak acids, giving up a proton only to very strong bases, such as sodium hydride or sodium amide.

Sulfones are generally unreactive compounds and are not attacked by most reducing agents. The bond joining the sulfur atom to an aromatic ring can be cleaved by piperidine in the presence of sodium amide:



alkyl aryl sulfone — piperidine — sodium amide

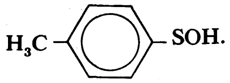
N-phenylpiperidine — sulfinate (salt) — ammonia

If two or three sulfonyl groups are bonded to the same carbon atom, a hydrogen atom bonded to that atom shows somewhat acidic properties; that is, it can be removed by a base.

Dimethyl sulfoxide has been investigated as a topical analgesic and as a carrier for drugs (it rapidly penetrates the skin); the disulfones sulfonal, trional, and tetronal have been used in medicine as hypnotics but have been replaced by less toxic compounds.

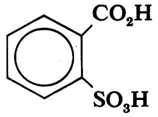
Sulfonic, sulfinic, and sulfenic acids. Three sulfur-containing functional groups confer acidity upon compounds in which they are present; these compounds are the sulfonic acids, in which the group SO_3H is present; the sulfinic acids, with the group SO_2H; and the sulfenic acids, with the group SOH. All three types of compounds are named by attaching the name of the functional group to the name of the compound in which that group replaces a hydrogen atom:

Cl_3CSO_3H $C_2H_5SO_2H$

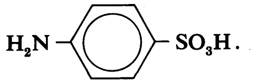


trichloromethanesulfonic acid — ethanesulfinic acid — *p*-toluenesulfenic acid

Occasionally, sulfonic acids are named using the prefix sulfo- or by arbitrary names:



o-sulfobenzoic acid



sulfanilic acid

The sulfonic acids are very strong—comparable to the mineral acids, hydrochloric, nitric, or sulfuric—and are the most common of the sulfur-containing acids. Most of them are colourless, odourless, crystalline compounds; the characteristic water solubility of these acids and most of their salts has made them useful as detergents that perform well in hard water and as dyes that can be applied to textiles from aqueous solutions.

Water solubility

Sulfinic acids are weaker, less soluble in water, and less stable than sulfonic acids; they are most often prepared in the form of their metal salts, which are more stable. Sulfenic acids and their salts are unstable compounds, rarely isolated; many substances named as derivatives of sulfenic acids are not actually obtainable from them.

Preparation

Sulfonic acids. Aromatic sulfonic acids usually are made by the reaction (sulfonation) of an aromatic hydrocarbon with sulfuric acid. Aliphatic hydrocarbons seldom react similarly with sulfuric acid, but aliphatic sulfonic acids may be obtained by oxidation of thiols or other sulfur-containing starting materials or by treatment of certain halogen compounds with sodium sulfite:

$$C_6H_6 + H_2SO_4 \longrightarrow C_6H_5SO_3H + H_2O$$

benzene	sulfuric acid	benzenesulfonic acid	water

$$RSH + 6HNO_3 \longrightarrow RSO_3H + 6NO_2 + 3H_2O$$

thiol	nitric acid	sulfonic acid	nitrogen dioxide	water

$$ClCH_2R + Na_2SO_3 \longrightarrow NaO_3SCH_2R + NaCl.$$

alkyl chloride	sodium sulfite	sodium alkanesulfonate	sodium chloride

The formation of sulfonic acids from sulfuric acid and aromatic compounds can be reversed by reaction with water, although strenuous conditions are sometimes necessary (*e.g.*, use of superheated steam). When aromatic sulfonic acids are heated with caustic alkalies, the sulfo ($-SO_3H$) group is replaced by a hydroxyl ($-OH$) group; this reaction is useful for preparing phenols:

$$C_6H_5SO_3H + NaOH \longrightarrow C_6H_5OH + NaHSO_3.$$

benzenesulfonic acid	sodium hydroxide	phenol	sodium bisulfite

In certain cases, the sulfo group can be replaced by various others, including cyano ($-CN$), nitro ($-NO_2$), formyl ($-CHO$), amino ($-NH_2$), or alkylthio ($-SR$).

Sulfonyl chlorides may be prepared from sulfonic acids by reaction with phosphorus pentachloride, chlorosulfuric acid, or certain other reagents:

$$C_6H_5SO_3H + PCl_5 \longrightarrow C_6H_5SO_2Cl + POCl_3 + HCl.$$

benzenesulfonic acid	phosphorus pentachloride	benzenesulfonyl chloride	phosphoryl chloride	hydrogen chloride

Sulfonyl chlorides are widely employed in preparing sulfonamides, sulfonates (esters of sulfonic acids), and sulfinic acids; their use in making sulfones has already been mentioned. Other sulfonyl halides (fluorides, bromides, or iodides) are less frequently made.

Sulfonamides result from the reaction of sulfonyl chlorides with ammonia or primary or secondary amines, compounds that all have the structure R_2NH, in which R represents a hydrogen atom or an organic group:

$$R'SO_2Cl + 2R_2NH \longrightarrow R'SO_2NR_2 + R_2NH_2^+Cl^-.$$

sulfonyl chloride	**primary or secondary amine or ammonia**	**sulfonamide**	**hydrochloride**

Sulfonates are obtained by treating alcohols or phenols with sulfonyl chlorides in the presence of a tertiary amine:

$$RSO_2Cl + R'{-}OH + R''_3N \longrightarrow RSO_2OR' + R''_3NH^+Cl^-.$$

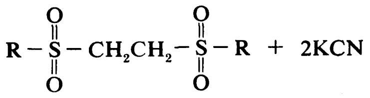
sulfonyl chloride	**alcohol**	**tertiary amine**	**sulfonate ester**	**amine hydrochloride**

Sulfinic acids. Sulfinic acids usually are made by treating a sulfonyl chloride with finely divided zinc or with sodium sulfite:

$$2RSO_2Cl + 2Zn \longrightarrow (RSO_2)_2Zn + ZnCl_2.$$

sulfonyl chloride	zinc	sulfinic acid (zinc salt)	zinc chloride

Aliphatic sulfinic acids can be made by this reaction, but the necessary sulfonyl chlorides are not always available; in such cases, the reaction of ethylene disulfones with potassium cyanide can be employed:



$$R-\overset{\overset{O}{\|}}{\underset{\underset{O}{\|}}{S}}-CH_2CH_2-\overset{\overset{O}{\|}}{\underset{\underset{O}{\|}}{S}}-R + 2KCN$$

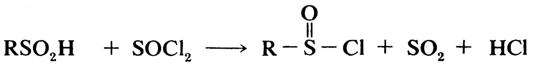
ethylene disulfone	potassium cyanide

$$\longrightarrow 2RSO_2K + N\equiv CCH_2CH_2C\equiv N.$$

potassium sulfinate	succinonitrile

Several other reactions have been utilized for making sulfinic acids, but they do not have general applicability. One of these reactions is the basis of the manufacture of the sodium salt of hydroxymethanesulfinic acid, sometimes called sodium formaldehydesulfoxylate, used in stripping dyes from textiles and in the discharge printing process for producing dyed designs on fabrics.

Sulfinic acids are oxidized to sulfonic acids by hydrogen peroxide or by nitric acid but to sulfonyl halides by chlorine, bromine, or iodine. The reaction of a sulfinic acid with thionyl chloride can be used to prepare sulfinyl chlorides, which are useful in the synthesis of sulfinamides and sulfinic esters (sulfinates):



$$RSO_2H + SOCl_2 \longrightarrow R-\overset{\overset{O}{\|}}{S}-Cl + SO_2 + HCl$$

sulfinic acid	thionyl chloride	sulfinyl chloride	sulfur dioxide	hydrogen chloride

Sulfenic acids. Sulfenic acids or their salts are formed as the initial products in reactions in which the sulfur-sulfur bond of disulfides is broken by the attack of alkalies; the molecules of sulfenic acids react rapidly with one another, producing sulfinic acids and thiols:

$$2RSOK \longrightarrow RSO_2K + RSK.$$

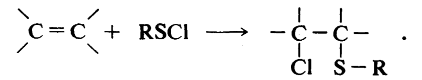
sulfenate (salt)	sulfinate (salt)	thiol (salt)

The sulfenyl chlorides are more stable than the sulfenic acids; they are formed in the reaction of chlorine with disulfides (see above *Disulfides and polysulfides*). The chlorides can be converted into other derivatives of sulfenic acids, such as amides, esters, or anhydrides:

$$2RSCl + H_2O \longrightarrow RSOSR + 2HCl.$$

sulfenyl chloride	water	sulfenic anhydride	hydrogen chloride

They also react with olefins to produce chlorine-containing sulfides:



olefin	sulfenyl chloride	chloroalkyl sulfide

LESS COMMON ORGANIC SULFUR COMPOUNDS

Organic derivatives of inorganic sulfur-containing acids. Several inorganic acids contain one or more atoms of sulfur in their molecules. The most important of these are sulfuric acid, H_2SO_4, and sulfurous acid, H_2SO_3. Most other inorganic acids contain one or more oxygen atoms that can be replaced by sulfur atoms. Nearly all of these acids have organic derivatives, such as esters or amides.

Sulfuric acid. Several esters of sulfuric acid are important industrial chemicals; dimethyl sulfate and diethyl sulfate, made from the alcohols and oleum (a solution of sulfur trioxide in sulfuric acid), are used to introduce methyl and ethyl groups into organic molecules. Certain monoesters of sulfuric acid occur as water-soluble forms in which substances are eliminated from the body.

Thiosulfuric acid. Monoesters of thiosulfuric acid, called Bunte salts, are prepared as intermediates in the synthesis of thiols:

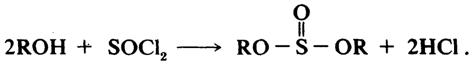
$$RBr + Na_2S_2O_3 \longrightarrow RSSO_3Na + NaBr$$

alkyl bromide	sodium thiosulfate	Bunte salt	sodium bromide

$$RSSO_3Na + NaOH \longrightarrow R-SH + Na_2SO_3.$$

Bunte salt	sodium hydroxide	thiol	sodium sulfite

Sulfurous acid. Esters of sulfurous acid can be made from alcohols and thionyl chloride:

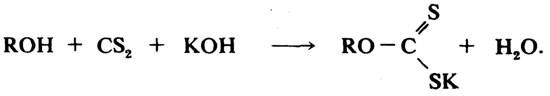


$$2ROH + SOCl_2 \longrightarrow RO-\overset{\overset{O}{\|}}{S}-OR + 2HCl.$$

alcohol	thionyl chloride	dialkyl sulfite	hydrogen chloride

Xanthates, rayon, and cellophane

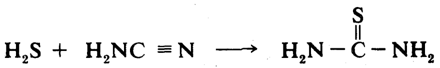
Carbonic acid. The derivatives of carbonic acid, H_2CO_3, have counterparts in which any or all of the oxygen atoms have been replaced by sulfur atoms. One of the important groups of these compounds is that of the xanthates (Greek *xanthos*, "yellow," from the colour of their copper salts), which are made from hydroxyl compounds and carbon disulfide:



alcohol	carbon disulfide	potassium hydroxide	xanthate (salt)	water

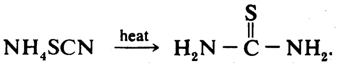
This reaction is used to produce a soluble form of cellulose that can be extruded into an acidic solution, which disrupts the xanthate group, regenerating the cellulose in the form of fibres (rayon) or films (cellophane). Xanthates of simpler alcohols are used as collectors in ore flotation (that is, agents that preferentially attach themselves to the surface of certain minerals, making air bubbles cling to them so that they float to the surface).

Thiourea, the diamide of thiocarbonic acid, is manufactured from hydrogen sulfide and cyanamide or by heating ammonium thiocyanate:



$$H_2S + H_2NC\equiv N \longrightarrow H_2N-\overset{\overset{S}{\|}}{C}-NH_2$$

hydrogen sulfide	cyanamide	thiourea

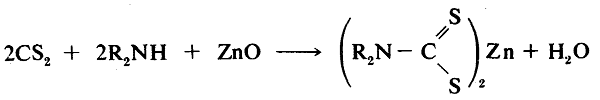


$$NH_4SCN \xrightarrow{heat} H_2N-\overset{\overset{S}{\|}}{C}-NH_2.$$

ammonium thiocyanate

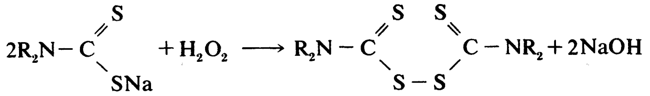
It is used as a component of photocopying papers and silver polishes and in a synthesis of thiols that prevents the formation of sulfides as by-products.

Several derivatives of dithiocarbamic acid are used as accelerators of the vulcanization of rubber. The following equations show typical methods of preparing these compounds:



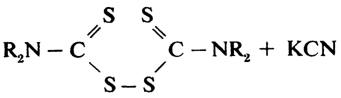
$$2CS_2 + 2R_2NH + ZnO \longrightarrow \left(R_2N-C(=S)S\right)_2Zn + H_2O$$

carbon disulfide · secondary amine · zinc oxide · zinc dialkyldithiocarbamate · water



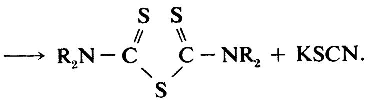
$$2R_2N-C(=S)SNa + H_2O_2 \longrightarrow R_2N-C(=S)-S-S-C(=S)-NR_2 + 2NaOH$$

sodium dialkyldithiocarbamate · hydrogen peroxide · tetraalkylthiuram disulfide · sodium hydroxide



$$R_2N-C(=S)-S-S-C(=S)-NR_2 + KCN$$

potassium cyanide



$$\longrightarrow R_2N-C(=S)-S-C(=S)-NR_2 + KSCN.$$

tetraalkylthiuram sulfide · potassium thiocyanate

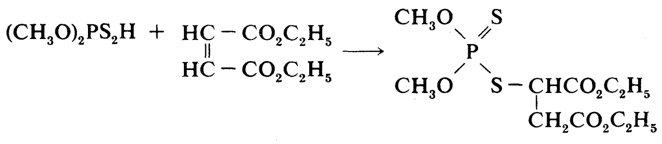
Phosphoric acid. Sulfur-containing derivatives of phosphoric acid, H_3PO_4, are useful insecticides, lubricant additives, and ore flotation agents. Some are made from tetraphosphorus dectasulfide or thiophosphoryl chloride:

$$P_4S_{10} + 8ROH \longrightarrow 4(RO)_2PS_2H + 2H_2S$$

tetraphosphorus decasulfide · alcohol · *O,O′*-dialkylphosphorodithioic acid · hydrogen sulfide

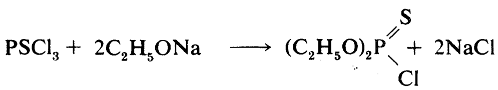
$$2(RO)_2PS_2H + ZnO \longrightarrow [(RO)_2PS_2]_2Zn + H_2O$$

zinc oxide · zinc *O,O′*-dialkylphosphorodithioate



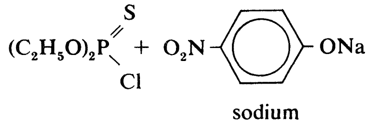
$$(CH_3O)_2PS_2H + \begin{matrix}HC-CO_2C_2H_5\\ \|\\ HC-CO_2C_2H_5\end{matrix} \longrightarrow (CH_3O)_2P(=S)-S-\begin{matrix}CHCO_2C_2H_5\\ |\\ CH_2CO_2C_2H_5\end{matrix}$$

O,O′-dimethylphosphorodithioic acid · diethyl maleate · Malathion

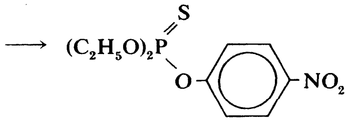


$$PSCl_3 + 2C_2H_5ONa \longrightarrow (C_2H_5O)_2P(=S)Cl + 2NaCl$$

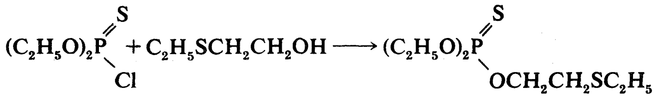
thiophosphoryl chloride · sodium ethoxide · *O,O′*-diethyl phosphorochloridothioate



sodium *p*-nitrophenoxide



parathion

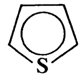


$$(C_2H_5O)_2P(=S)Cl + C_2H_5SCH_2CH_2OH \longrightarrow (C_2H_5O)_2P(=S)OCH_2CH_2SC_2H_5$$

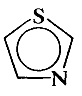
2-ethylthioethanol · Systox

Heterocyclic sulfur compounds. In several groups of organic compounds, sulfur atoms are present in a ring structure (see below *Heterocyclic compounds*). The most important of these cyclic compounds have the following molecular formulas:

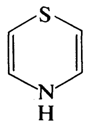
Ring structure



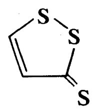
thiophene



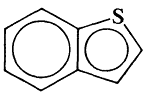
thiazole



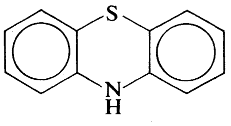
1,4-thiazine



1,2-dithiole-3-thione



thianaphthene or benzothiophene



phenothiazine

(J.V.K.)

Heterocyclic compounds

Heterocyclic compounds are a major class of organic chemical compounds, characterized by the fact that the atoms in their molecules are joined into rings, or circles, containing at least one atom of an element other than carbon. These compounds are of great importance because many of the biochemical materials essential to life belong to the class. Nucleic acids, for example, the chemical substances that carry the genetic information controlling inheritance, consist of long chains of heterocyclic units held together by other types of materials. Many naturally occurring pigments, vitamins, and antibiotics are heterocyclic compounds, as are most hallucinogens, substances that produce hallucinations. Modern society is dependent on synthetic heterocycles for use as pharmaceuticals, pesticides, and herbicides, as well as dyes and plastics.

The most common heterocycles are those with five- or six-membered rings, containing atoms of nitrogen, oxygen, or sulfur. The best known of the simple heterocyclic compounds are pyridine, pyrrole, furan, and thiophene. Pyridine and pyrrole are both nitrogen heterocycles; that is, their molecules contain nitrogen atoms in the rings with carbon. Because the molecules of many biological materials consist in part of pyridine and pyrrole rings, strong heating of such materials produces small amounts of pyridine and pyrrole. In fact, both of these substances were discovered in the 1850s in an oily mixture of substances formed by strong heating of bones. Today, however, pyridine and pyrrole are generally obtained from coal tar or are prepared by synthetic reactions. The chief commercial interest in both compounds lies in their conversion to other substances, chiefly dyestuffs and drugs. Pyridine is used also as a solvent, a waterproofing agent, a rubber additive, an alcohol denaturant, and a dyeing adjunct. Furan is an oxygen-containing heterocycle, used primarily for conversion to other substances (including pyrrole). Furfural, a close chemical relative of furan, is obtained from oat hulls and corncobs and is used in the production of ingredients of nylon. Thiophene, a sulfur heterocycle, resembles benzene in its chemical and physical properties. It is a frequent contaminant of benzene from natural sources and was first discovered during purification of benzene. Like the other compounds, it is used primarily for conversion to other substances. Furan and thiophene both were discovered in the latter part of the 19th century.

The common simple heterocyclic compounds

Each of the molecules of pyridine contains a ring of six atoms, five of which are carbon atoms and one of which is a nitrogen atom. Pyrrole, furan, and thiophene molecules contain five-membered rings, composed of four atoms of carbon and one atom of nitrogen, oxygen, or sulfur, respectively.

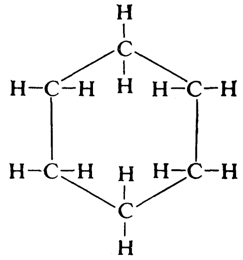
In general, the physical and chemical properties of heterocyclic compounds are best understood by comparing them with ordinary organic compounds that do not contain hetero-atoms (noncarbon atoms). In this section, the comparative aspects of heterocyclic compounds will be considered first, followed by a brief survey of the different classes of heterocyclic compounds, arranged according to the size and nature of the ring.

GENERAL ASPECTS OF HETEROCYCLIC SYSTEMS

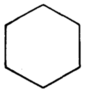
Comparison with carbocyclic compounds. The molecules of organic chemical compounds are built up from a framework or backbone of carbon (C) atoms to which are attached hydrogen (H), oxygen (O), or other heteroatoms. Carbon atoms have the unique property of being able to join with one another to form chains of atoms. When the ends of the chains are joined together, ring—that is, cyclic—compounds result; such substances often are referred to as alicyclic or carbocyclic compounds. Substitution of one or more of the ring carbon atoms in the molecules of a carbocyclic compound with a heteroatom gives a heterocyclic compound.

Carbocyclic compounds

A typical carbocyclic compound is cyclohexane (C_6H_{12}), the molecular structure of which is indicated by the following formula

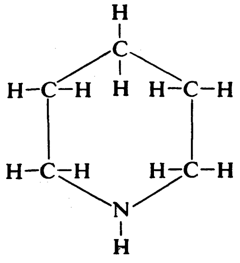


in which the letters represent atoms of the elements of which they are symbols and the lines represent bonds (linkages) between the atoms. For convenience such formulas are often written in the simplified polygonal form, such as

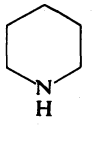


for cyclohexane, in which each angle of the polygon represents a carbon atom (it being understood that hydrogen atoms are joined to the carbon atoms as required).

When one of the carbon atoms of cyclohexane is replaced with an atom of nitrogen, the compound piperidine (a chemical relative of pyridine, above) is produced. The structural formula of piperidine is written as follows:



; or for convenience:



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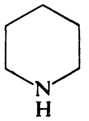
Other heterocyclic compounds can be envisioned as having been produced (similarly) from cyclohexane by substitution with other hetero-atoms or from other carbocyclic compounds by substitution with nitrogen or other hetero-atoms.

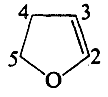
araf-
inic,
lefinic,
nd
romatic
om-
ounds

The simplest organic compounds are the hydrocarbons, compounds of carbon and hydrogen only. Hydrocarbons are classed as paraffinic if all the potential bonds of the carbon atoms are saturated—that is, if the four possible bonds of each carbon atom are joined singly to another carbon atom or to a hydrogen atom. They are classified as olefinic if they contain a double bond (also called an unsaturated linkage) between any two of the carbon atoms; and they are classed as aromatic if they contain a ring of alternating single and double bonds. Compounds with unsaturation are highly reactive—that is, they readily undergo additions of atoms or groups of atoms to the carbon atoms of their double bonds, giving each carbon four groups joined to it. Aromatic compounds, though unsaturated, are extremely stable and do not undergo the addition reactions characteristic of other unsaturated compounds. The stability and unreactivity of the aromatic system are associated with presence of three pairs of electrons, the so-called pi electrons, associated with the three double bonds of the ring. Together these electrons, comprising the so-called aromatic sextet, form an unusually stable structure, associated with the aromatic ring as a whole rather than with the individual atoms.

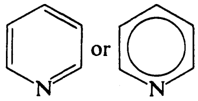
Heterocycles, too, may be classified as paraffinic, olefinic, and aromatic. Thus, piperidine is a heteroparaffinic compound containing no double bond, whereas 4,5-dihydrofuran is a hetero-olefinic compound, and pyridine is a typical hetero-aromatic substance, as shown in the following structural formulas, in the first of which the double bonds are shown and in the second of which the aromatic sextet is indicated by a circle.



piperidine

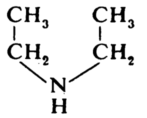


4,5-dihydrofuran

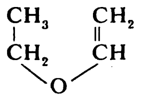


pyridine

This classification relates the chemistry of heterocycles directly with that of nonheterocyclic derivatives, which are usually better known. In general, synthetic methods and physical and chemical properties of the saturated heteroparaffinic and the partly unsaturated hetero-olefinic compounds closely resemble those for their acyclic (noncyclic) analogues. Thus, piperidine may be considered as a cyclic secondary amine (organic nitrogen compound) and has much in common with the acyclic amine, diethylamine, which is represented as follows:



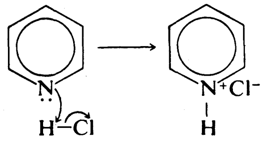
Similarly, 4,5-dihydrofuran mirrors many of the properties of the unsaturated ether, ethyl vinyl ether, written as follows:



It is within the area of hetero-aromatic compounds that most of the novel chemistry of the class is to be found, and for that reason hetero-aromatics will be emphasized in this section.

Chemical reactions. Every chemical reaction involves the forming or breaking of a chemical bond: for heterocycles the bonds in question are generally covalent bonds—linkages consisting of a pair of electrons shared by two atoms. Reactions generally can be placed in one of several categories, depending upon the origin, or disposition, of the electron pair of the bond that is formed or broken. The first of these categories is heterolytic reactions, in which the electron pair is supplied by one of the reactants (known as the nucleophile) to the other reactant (known as the electrophile). In the formation of pyridinium chloride, for example, which occurs as shown in the following equation:

Heterolytic reactions



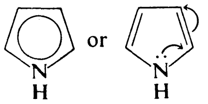
in which, as is customary, the structure of the starting material is written at the left, the product is written at the right, and an arrow in between the two indicates the reaction. In these structural formulas the presence of a free electron pair on the nitrogen atom is indicated by two dots, the movements of electrons are shown by curved arrows, and plus and minus signs indicate positive and negative charges. In this reaction, the pyridine molecule behaves as a nucleophile, the electron pair that initially resided on the nitrogen atom finally being shared between that atom and the hydrogen atom of the hydrogen chloride (which acts, therefore, as an electrophile). In homolytic reactions, an electron-pair bond is formed (or broken) and in this process each of the reagents donates (or receives) a single electron. In some reactions, cyclic transition states are involved, which must be classified separately.

In a heterolytic bond-forming reaction the reagent that accepts the electrons is known as an electrophilic reagent; for instance, in the coordination reaction with hydrogen chloride (above), the nitrogen atom of pyridine acts as a nucleophilic reagent. This behaviour is characteristic of pyridine. The nitrogen atom in pyrrole, by contrast, does not act as a nucleophilic centre, a distinction that can be rationalized by the different type of electronic interaction present in these two hetero-aromatic molecules (as explained below).

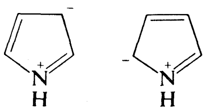
The nature of hetero-aromaticity. Aromaticity denotes the significant stabilization of a ring compound by cyclic conjugation in which six pi electrons generally participate. A nitrogen atom in a ring can carry a positive or negative charge, or it can be in the neutral form. Ring oxygen and sulfur atoms can occur either in the neutral form or with a positive charge. A fundamental distinction is usually made between: (1) those hetero-atoms that enter into aromatic conjugation by means of a lone electron pair in an orbital perpendicular to the plane of the ring, and (2) those hetero-atoms that enter into conjugation because they are connected to another atom by means of a double bond.

Pyrrole- and pyridine-type hetero-atoms

An example of an atom of the first type is the nitrogen atom in pyrrole. In pyrrole, the aromatic electron sextet is made up by the participation of two electrons from each carbon–carbon double bond, as well as the two electrons that comprise the unshared pair of the nitrogen atom. As a consequence, there tends to be a net flow of electron density from the nitrogen to the carbon atoms as the nitrogen electrons are drawn into the aromatic sextet; this movement may be indicated as follows,

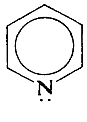


in which the curved arrows depict the movement, or tendency toward movement, of the electron pair. Alternatively, the pyrrole molecule may be described as a resonance hybrid—that is, a molecule whose true structure can only be approximated by two or more different forms, called resonance forms. With pyrrole, resonance forms such as the following make important contributions to the overall structure.

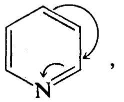


In these forms the nitrogen atom is depleted of electrons, as indicated by the positive charge.

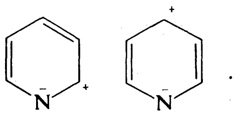
An example of a hetero-atom of the second type is the nitrogen atom in pyridine. Pyridine also has a pi electron sextet; but the nitrogen atom contributes only one electron to it, one further electron being contributed by each of the five carbon atoms. In particular, the lone electron pair on the nitrogen atom is not involved, as indicated in the following formulation:



Moreover, because of the greater attraction for electrons (electronegativity) of the nitrogen atom as compared to carbon atoms, there is a tendency for an electron flow toward the nitrogen atom, rather than away from it, as in pyrrole. This movement of electrons is shown as below

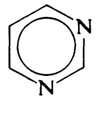


or it can be expressed in terms of contributions of the resonance forms as follows:

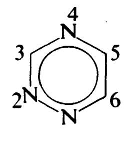


Quite generally, hetero-atoms may be referred to as pyrrole-like or pyridine-like, depending upon whether they fall into the first or second class, respectively, above. The pyrrole-like hetero-atoms —NR— (R being a hydrocarbon group), $-N^{-}-$, —O—, and —S— tend to donate electrons into the pi electron system, whereas the pyridine-like hetero-atoms —N=, $-N^{+}R=$, $-O^{+}=$, and $-S^{+}=$ tend to attract the pi electrons of a double bond.

A six-membered hetero-aromatic ring contains one or more pyridine-like hetero-atoms (usually nitrogen), as is the case with the compounds pyrimidine and 1,2,4-triazine.

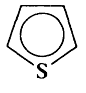


pyrimidine

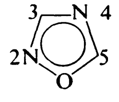


1,2,4-triazine

A six-membered hetero-aromatic compound cannot normally contain any pyrrole-like hetero-atoms. A five-membered hetero-aromatic ring, however, always contains one pyrrole-like nitrogen, oxygen, or sulfur atom, and it may also contain up to four pyridine-like hetero-atoms, as shown by the compounds below.



thiophene



1,2,4-oxadiazole

Measurements of aromaticity

The quantitative measurement, or even the precise definition, of aromaticity is difficult; two methods have been widely used to measure the aromaticity of carbocyclic compounds. The first method depends on the determination of the heat given off on complete combustion of the aromatic compound—say, benzene—compared with that expected for the corresponding carbocyclic compound containing three conventional double bonds—say, a cyclohexatriene. Because the heat of combustion is related to the energy content of the molecule, comparisons of the above kind reveal the reduction of energy associated with the aromatic system. The second method depends on measurement of magnetic ring currents set up in aromatic systems by external magnetic fields. The measurement is made with a nuclear magnetic resonance spectrometer, and the degree of observed magnetic effect is related to the degree of aromaticity. Both methods are difficult to apply to hetero-aromatic systems because of complications

Table 34: Aromatic Stabilization Energies for Benzene and Some Hetero-aromatic Rings
(kcal/mol)

benzene	pyridine	pyrazine
36	32	24
pyrrole	thiophene	furan
14–31	24–31	17–23
pyrazole	imidazole	1,2,4-triazole
27–41	12–32	20–49

arising from the presence of the heteroatoms. Table 34 gives some values for aromatic stabilization determined by combustion or related methods. Unfortunately, high errors are involved, as indicated by the wide variations in some of the values.

With regard to its chemical reactivity, an aromatic compound is characterized by extra stability of the conjugated system it contains; this characteristic, in turn, is denoted by the tendency of the compound to react by substitution (replacement of a hydrogen atom) rather than by addition to the double bonds. From a reactivity standpoint, therefore, the degree of aromaticity is measured by the relative tendency toward substitution rather than addition. Judging by this criterion, pyridine is more aromatic than furan, but it is difficult to say how much more aromatic it is.

Physical properties. Physical properties are important as criteria for judging the purity of heterocycles, just as for other organic compounds. The melting point was once the criterion most widely used, but now the optical spectra (based on light absorption), mass spectra (based on relative masses), and magnetic-resonance spectra (based on nuclear properties) have been increasingly used. Nevertheless, knowledge of the boiling point or melting point is still helpful in following purification of a compound. Organic compounds generally show great regularity regarding their physical properties, and heterocycles are no exception.

Criteria for purity

Table 35: Boiling Points of Saturated Heterocycles and Carbocycles of the Same Ring Size
(°C at 1 atm)

ring size	number and orientation of hetero-atoms	hetero-atom type			saturated cycloalkane
		NH	O	S	
3	one	56	11	55	−34
4	one	63	48	94	13
5	one	87	66	121	49
6	one	106	88	141	80
6	two (1, 2)	165*	—	—	80
6	two (1, 3)	150	106	—	80
6	two (1, 4)	145	101	200	80
7	one	138	120	174	119

*Corrected to atmospheric pressure by the method of Hass and Newton, *Handbook of Chemistry and Physics*, 51st ed., edited by R.C. Weast, 1970.

Melting and boiling points. The boiling points of certain saturated heterocycles compared with the corresponding cycloalkanes are listed in Table 35. The melting or boiling points of common hetero-aromatic ring compounds and their substituted derivatives are compared with those for benzene and its derivatives in Table 36. It can be seen from the table that replacement of a two-carbon unit (with molecule weight equal to 26) by a sulfur atom (atomic weight 32) has little effect on the melting or boiling point (compare corresponding benzene and thiophene compounds). On the other hand, replacement of a two-carbon unit by an oxygen atom (atomic weight 16) lowers the boiling point by about 40° C, or 104° F (compare corresponding benzene and furan compounds), which is to be expected because of the decreased molecular weight (lighter compounds being more volatile). The introduction of nitrogen atoms into the benzene ring is accompanied by less regular changes. Replacement of a two-carbon unit by an imino (NH) group, or of a single carbon by a nitrogen atom, increases the boiling point. Furthermore, if these two changes are made simultaneously, the boiling point is increased by an especially large amount, probably the result of association by hydrogen bonding (a weak form of attachment through certain types of hydrogen atoms) between the nitrogen atom and the imino group.

Effects of substituents

The effects of substituent groups in hetero-aromatic rings show considerable regularity. Thus, methyl and ethyl groups attached to ring carbon atoms usually increase the boiling point by about 20°–30° C (68°–86° F) and 50°–60° C (122°–140° F), respectively, whereas conversion of an imino (NH) group into an amino (NR, in which R is a hydrocarbon chain) group (*e.g.,* pyrazole → 1-methylpyrazole) significantly decreases the boiling point because of decreased ease of association by hydrogen bonding (the active hydrogen having been replaced by a hydrocarbon group). Carboxylic acids and amides are all solids; carboxy derivatives of compounds containing a ring nitrogen atom usually melt at higher temperatures than those containing ring oxygen or sulfur atoms because of hydrogen bonding. Compounds containing both a ring nitrogen atom and a hydroxyl or amino group are usually relatively high melting solids. Chloro compounds usually have boiling points similar to those of the corresponding ethyl compounds.

Ultraviolet, infrared, nuclear magnetic resonance and mass spectra. Spectroscopic studies of heterocyclic compounds, as have those of other organic compounds, have become of great importance as means of identification of unknown materials, as criteria of purity, and as probes for investigating the electronic structures of molecules, thereby explaining and helping to predict their reactions. The pattern of light absorption in the ultraviolet region of the spectrum (the so-called ultraviolet spectrum) of a compound is characteristic of the pi electron system of the molecule; *i.e.,* of the arrangement of double bonds within the structure. The ultraviolet spectra of hetero-aromatic

Table 36: Melting and Boiling Points of Hetero-aromatic Compounds*
(°C at 1 atm)

ring system (with location of substituent)	substituent										
	H	CH_3	C_2H_5	CO_2H	$CO_2C_2H_5$	$CONH_2$	NH_2	OH	OCH_3	Cl	Br
Benzene	80	111	136	122†	211	130†	184	43†	37†	131	155
Pyridine-2	115	128	148	137†	243	107†	57†	107†	252	171	193
Pyridine-3	115	144	163	235†	223	129†	65†	125†	179‡	150	173
Pyridine-4	115	145	171	306†	219	156†	157†	148†	93†	147	174
Pyrrole-1	130	114	129	95†‡	180	166†	—	—	—	—	—
Pyrrole-2	130	148	181	205†‡	39†	174†	—	—	—	—	—
Pyrrole-3	130	158	179	148†	78†‡	152†	—	—	—	—	—
Furan-2	31	64	92	133†	34†	142†	68†	80†	110	78	103
Furan-3	31	65	—	122†	179	168†	—	58†	—	80	103
Thiophene-2	84	113	133	129†	218	180†	214	217	154	128	150
Thiophene-3	84	115	135	138†	208	178†	—	—	—	136	157
Pyrazole-1	70†	127	137	—	213†	—	—	—	—	—	—
Pyrazole-3	70	205	209	214†‡	160†	—	285	164†	—	—	—
Pyrazole-4	70	207	—	275†	—	—	81†	118†	—	77†	97†
Isoxazole-3	95	118	138	149†‡	—	134†	—	—	—	—	—
Isoxazole-5	95	122	—	149†	—	174†	—	—	—	—	—
Imidazole-1	90	199	226	—	218†	—	—	—	—	—	—
Imidazole-2	90	141†	80	164†‡	—	—	—	250†‡	—	—	207†
Imidazole-4	90	56†	—	275†‡	157†	215†	—	—	—	—	130
Pyrimidine-2	123	138	—	270†	—	—	127†	320†	—	65†	—
Pyrimidine-4	123	141	—	240†‡	—	—	151†	164†	—	—	—
Pyrimidine-5	123	153	—	270†	38†	212†	170†	210†‡	—	—	75†
Pyrazine-2	57†	135	—	229†‡	—	189†	—	119†	187†	160	180

*Melting points above 30° C are indicated by a dagger†; those below 30° C are not included.
A dash indicates the compound is unstable, unknown, or that the data are not readily available.
‡Indicates that the compound melts with decomposition.

Table 37: Ultraviolet Spectral Characteristics of Hetero-aromatic Compounds

	neutral form		form with single positive charge			neutral form		form with single positive charge	
	wavelength (nm)	intensity ($\log_{10}\epsilon$)	wavelength (nm)	intensity ($\log_{10}\epsilon$)		wavelength (nm)	intensity ($\log_{10}\epsilon$)	wavelength (nm)	intensity ($\log_{10}\epsilon$)
Pyrrole	210	4.20	241	3.90	Pyridine	257	3.42	256	3.70
Furan	208	3.90	—	—	Pyridazine	247	3.04	238	3.21
						300	2.51		
Thiophene	231	3.87	—	—	Pyrimidine	243	3.51	242	3.64
Pyrazole	210	3.53	217	3.67	Pyrazine	261	3.77	266	3.86
						300	2.93		
Isoxazole	211	3.60	—	—	Indole	270	3.77	280	3.68
					Quinoline	275	3.51		
Isothiazole	244	3.72	—	—		299	3.46	313	3.79
						212	3.52		
Thiazole	233	3.57	—	—	Isoquinoline	306	3.38	270	3.30
						319	3.47	332	3.63

Measurements of pi electrons

compounds (given in Table 37) show general similarity to those of benzenoid compounds, and the effects of substituents can usually be rationalized in a similar way. The infrared spectrum of an organic compound, with its complexity of bands, provides an excellent "fingerprint" of the compound (far more characteristic than the melting point), and it can also be used to identify certain common groups, such as carbonyl (C=O) and imino (N—H) groups, as well as various of the heterocyclic ring systems (Table 38). Although magnetic resonance spectra were not widely used until about 1955, they have since become indispensable to any serious study of heterocyclic chemistry. Proton resonance spectra (the most common type), for example, yield information regarding the number of hydrogen atoms present in the molecule, as well as their chemical environment and their relative orientations in space (Table 39). Still later, mass spectra have been used to determine not only the complete molecular formula of the compound but also, from the way the molecule is fragmented, the arrangement of many of the atoms.

Table 38: Ranges for Some Characteristic Infrared Absorption Maxima for Ring-Stretching Modes of Common Hetero-aromatic Compounds
(cm^{-1})

compound	absorption frequencies		
Pyrroles	1560–1530	1510–1480	1410–1390
Furans	1600–1560	1520–1470	1410–1370
Thiophenes	1540–1505	1440–1405	1370–1340
Pyridines	1610–1590	1580–1570	1485–1465

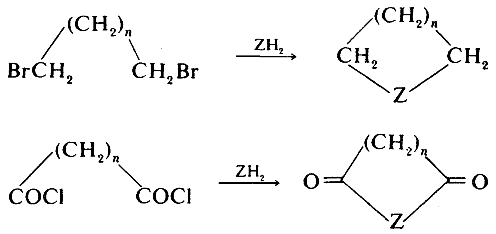
Table 39: Chemical Shifts in Proton Resonance Spectra of Various Heterocycles
(parts per million on a delta scale)

compound	chemical shift at position		
	2	3	4
Aziridine	1.48	—	—
Oxirane	2.54	—	—
Thiirane	2.27	—	—
Azetidine	3.58	2.32	—
Pyrrole	6.62	6.05	—
Furan	7.40	6.30	—
Thiophene	7.19	7.04	—
Pyridine	8.50	7.06	7.46
Pyridazine	—	9.17	7.68

Synthesis and modification of heterocyclic rings. The important methods for synthesizing heterocyclic compounds may be classified under five headings, of which the first three are ways of forming new heterocyclic rings from precursors containing one less ring, the fourth covers the formation of a heterocyclic ring either from another heterocycle or from a carbocyclic ring, and the fifth includes the modification of substituents on a pre-existing heterocyclic ring. The key step in the formation of rings from noncyclic precursors (the first instance above) often is the formation of the carbon hetero-atom linkage (C—Z, in which Z represents an atom of nitrogen, oxygen, or sulfur). The actual ring closure, or cyclization, however, may involve the formation of a carbon–carbon bond. In any case, ring formation reactions are subdivided into three categories according to whether the cyclization reaction occurs primarily as a result of (1) nucleophilic or (2) electrophilic attack, or (3) by way of a cyclic transition state.

Classification by methods of ring closure

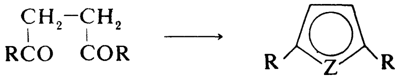
Nucleophilic ring closure. To prepare compounds containing one hetero-atom, a compound containing two halides (chloride [Cl], bromide [Br], or iodide [I]) or acyl halides (that is, halide derivatives of carboxylic acids) will react with the dihydro form of the hetero-atom, ZH_2 (or an equivalent reagent such as ammonia or sodium sulfide), to give nonaromatic heterocycles:



in which n equals only an integer

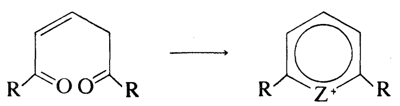
Diketones

Diketones (ketones being compounds with oxygen atoms joined to carbon by double bonds) also can react with dihydro-Z compounds to give heterocycles. Diketones with the keto groups separated by two carbon atoms, for example, can be cyclized to form aromatic pyrroles, furans, and thiophenes as shown:



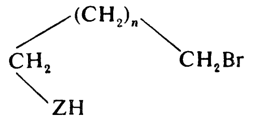
in which R is a hydrocarbon chain

With diketones separated by three carbons, such as the unsaturated compound below, six-membered rings may be formed:

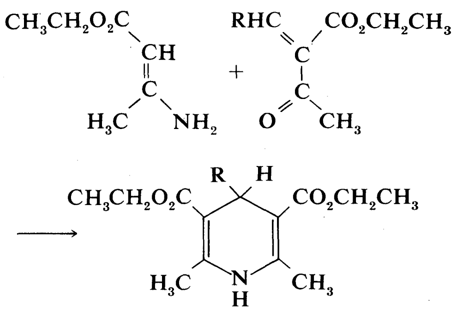


In each of these reactions the hetero-atom Z acts as a nucleophile in attacking the positively charged carbon atom produced by electron withdrawal due to the halogen atom (in the first instance above) or to the oxygen atom (in the second).

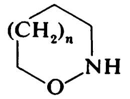
Usually, reactions like these proceed by means of intermediates in which only one of the C—Z bonds has been formed. In reactions of the first type in this section, for instance, compounds like the one shown below may be formed first.



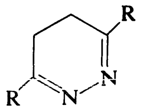
It is frequently possible to synthesize such intermediate compounds by other routes, and they then cyclize readily. One form of pyridine synthesis, for example, involves the condensation of an intermediate with the carbon–nitrogen bond already formed as shown:



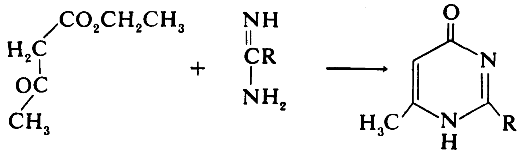
Heterocycles containing two adjacent nitrogen atoms and oxygen atoms, or both, also may be prepared from precursors of the type shown in the first example of this section by using hydrazine (N_2H_4), hydroxylamine (NH_2OH), or hydrogen peroxide (H_2O_2) in place of the dihydro-Z compound (ZH_2). With hydroxylamine, for example, the compound shown below is formed.



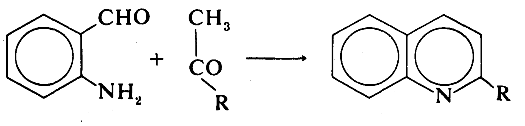
Similarly, two adjacent hetero-atoms can be introduced by reactions with diketones (the second reaction here), producing such compounds as the following:

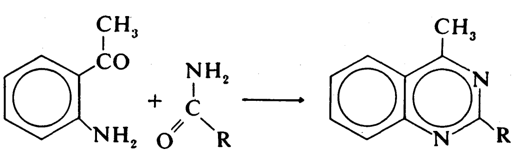


When it is desired to prepare a compound containing two nonadjacent hetero-atoms, appropriate components can be put together as illustrated in the following synthesis of pyrimidines:

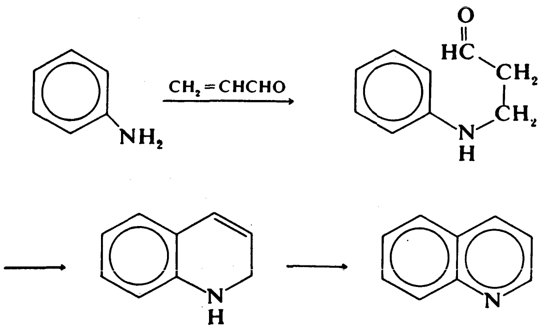


Ring synthesis reactions, in which the hetero-atom acts as a nucleophile, also involve orthodisubstituted benzenes (ortho substituents being on adjacent carbon atoms in the benzene ring), as in the preparation of many heterocycles fused to benzene rings. The formation of quinoline and quinazoline rings are examples of this type:



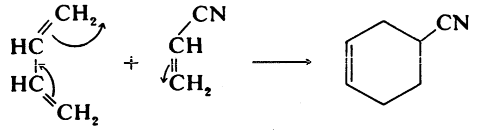


Electrophilic ring closure. Formation of heterocyclic rings by reactions in which the hetero-atom acts as an electrophilic (electron-seeking) reagent are rare, because the nitrogen, oxygen, and sulfur atoms are themselves electron-rich centres that react generally as nucleophiles. Electrophilic ring-closure reactions are known, however, in which a heterocyclic ring is formed by a reaction in which a carbon atom of the ring acts as an electrophile. Usually such reactions involve ring closure onto a benzene ring (or other aromatic system), an electron-rich system that is generally subject to attack by electrophilic reagents. An example of ring closures of this type is the formation of quinoline from aniline and acrolein, a dehydration product of glycerol. As shown below, the initial product of the reaction is a dihydroquinoline, which must be dehydrogenated to the fully aromatic product, quinoline itself:

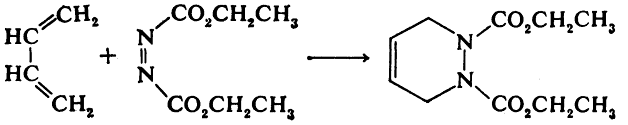


Ring closure by way of cyclic transition states. A most important method for the synthesis of carbocyclic six-membered rings is the so-called Diels–Alder diene reaction (named for its discoverers, two German Nobel Prize-winning chemists, Otto Diels and Kurt Alder). In this reaction, as illustrated below, a diene (a compound with two double bonds) reacts with a dienophile (a diene-seeking reagent) to yield a cyclohexene:

The Diels–Alder reaction



In this reaction, it is not possible to determine whether the electrons move clockwise (as shown by the curved arrows) or whether they move in the reverse, counterclockwise fashion. For this reason, in reactions of this type the movements of electrons are thought to be synchronous rather than sequential, and the reactions are said to possess cyclic transition states. Heterocycles can also be prepared by the diene synthesis, as illustrated by this preparation of a tetrahydropyridazine:



Of far greater use, however, is the related method referred to as the Huisgen cyclic dipolar addition reaction; it is an important method for the preparation of many types of five-membered rings, especially those containing several hetero-atoms. Pyrazoles and isoxazoles, and many less common heterocycles, also are prepared by this method.

Conversion of one heterocyclic ring into another. There are many reactions of theoretical, and a few of practical, importance in which one heterocyclic ring is converted into another. The ring-atom exchange reactions of pyridine derivatives (see below, *Six-membered rings with one hetero-atom*) are good examples. In addition, ring-atom rearrangement or "shuffling" can be brought about with light (photochemically) in five- and six-membered hetero-aromatic compounds, and ring contraction by extrusion of an atom or group can occur under certain conditions.

Modification of an existing ring. Dehydrogenation of saturated or partially saturated rings to hetero-aromatic compounds by heating with sulfur or treatment with a palladium catalyst is analogous to similar reactions with carbocyclic compounds. The hydrogenation of heteroaromatic rings is, by contrast, usually more difficult, for the hetero-atoms tend to poison the catalyst. Finally, the modification of substituents on heterocyclic rings is of highest importance in synthesis; reactions by which substituents may be altered are among the most useful in heterocyclic chemistry.

Common and systematic nomenclature

Nomenclature. Naming heterocyclic compounds is complicated because of the existence of many common names, in addition to the internationally agreed systematic nomenclature. A brief account of systematic nomenclature is given here, but for common names the reader is referred to the systematic survey in the next section.

The types of hetero-atoms present in a ring are indicated by prefixes; in particular "oxa," "thia," and "aza" denote oxygen, sulfur, and nitrogen atoms, respectively. The number of hetero-atoms of each kind are indicated by number prefixes joined to the hetero-atom prefixes as "dioxa" and "triaza." The presence of different heteroatoms is indicated by combining the above prefixes, using the following order of preference: oxa first, followed by thia, then aza. Ring size and the number of double bonds are indicated by suffixes as shown in Table 40.

Table 40: Suffixes Used in Naming Monocyclic Heterocycles

ring size	rings with nitrogen			rings without nitrogen		
	maximum unsaturated	one double bond	saturated	maximum unsaturated	one double bond	saturated
3	-irine	—	-iridine	-irene	—	-irane
4	-ete	-etine	-etidine	-ete	-etene	-etane
5	-ole	-oline	-olidine	-ole	-olene	-olane
6	-ine	—	—	-ine	—	-ane
7	-epine	—	—	-epine	—	-epane

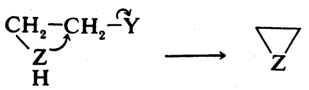
In this table, "maximum unsaturation" refers to rings possessing the largest possible number of noncumulative (that is, with no atom in common) double bonds but with oxygen and sulfur atoms in their normal, divalent forms. In addition, partially saturated rings are indicated by the prefixes "dihydro," "tetrahydro," and so on, according to the number of excess hydrogen atoms. The positions of hetero-atoms, extra hydrogen atoms, and substituents are indicated by arabic numerals, for which the numbering starts at an oxygen atom, if one is present, or at a sulfur or nitrogen atom, and continues in such a way that the hetero-atoms are assigned the lowest possible numbers. Other things being equal, numbering starts at a nitrogen atom that carries a substituent rather than at a multiply-bonded nitrogen. In compounds with maximum unsaturation, if the double bonds can be arranged in more than one way, their positions are defined by indicating the nitrogen or carbon atoms that are not multiply-bonded, and consequently carry an "extra" hydrogen atom (or substituent), as follows: 1H–, 2H–, and so on.

Numbering systems

MAJOR CLASSES OF HETEROCYCLIC COMPOUNDS

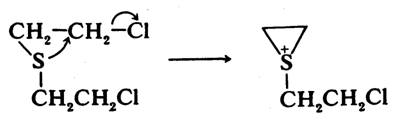
The major classes of heterocycles containing the common hetero-atoms (nitrogen, oxygen, sulfur) are reviewed first, in order of ring size (from smallest to largest), leaving compounds with other hetero-atoms to a final section. The classification by ring size is convenient because heterocyclic rings of a given size have many common features. For heterocyclic (as for carbocyclic) rings, the following broad generalizations can be made: 3- and 4-membered rings are strained and thus readily opened; they are also readily formed: such heterocycles are well-known reactive intermediates. Five- and six-membered rings are readily formed and are very stable: these sizes of ring also allow the development of aromatic character. Seven- and larger-membered rings are stable but not readily formed and are therefore relatively little investigated.

Three-membered rings. The three-membered-ring heterocycles containing single atoms of nitrogen, oxygen, and sulfur—aziridine, oxirane (or ethylene oxide), and thiirane, respectively—and their derivatives can all be prepared by nucleophilic reactions, of the type shown below.



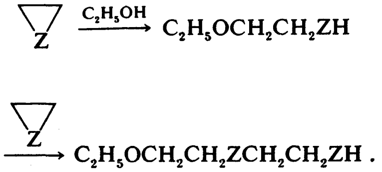
aziridine: Z=NH
oxirane: Z=O
thiirane: Z=S
(Y equals a group displaced by the heteroatom Z)

Thus, aziridine itself is formed by heating β-aminoethyl hydrogen sulfate with base (in which case Y is $-OSO_3H$). A reaction of this type is involved in the action of mustard gas, or dichlorodiethyl sulfide, a poison gas used in World War I. The mustard gas itself ringcloses to give a thiirane derivative, the biologically active agent, according to the reaction below:



Commercially, oxirane and aziridine (to a lesser extent) are important bulk industrial chemicals; oxirane is prepared on a large scale by the direct reaction of ethylene with oxygen.

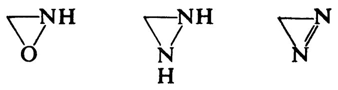
The chemical reaction characteristic of these three-membered rings is susceptibility to attack by nucleophilic reagents to open the ring as shown below:



The first step shown is the reverse of the formation reaction above. As indicated, a second molecule of the three-membered ring may react with the first product. Further reaction leads to long, chainlike molecules (polymers) of the type $C_2H_5O(CH_2CH_2Z)_nCH_2CH_2ZH$. Polymers and copolymers of oxirane and aziridine are useful plastics. Although aziridine, oxirane, and thiirane are heteroparaffinic, they are much more reactive than the corresponding normal amines, ethers, or sulfides because of the strain inherent in the three-membered ring. This behaviour reflects the comparable enhanced reactivity of the related carbocyclic compound cyclopropane.

Three-membered rings with two hetero-atoms

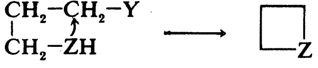
Since 1950, three different classes of three-membered-ring compounds with two hetero-atoms have been discovered. They are derivatives of the parent ring systems oxaziridine, diaziridine, and diazirine, the chemical structures of which are shown below:



oxaziridine diaziridine diazirine

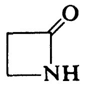
These compounds have no practical importance, but they are stable enough for easy handling and offer interesting possibilities for future use.

Four-membered rings. Azetidine, oxetane, and thietane—the four-membered rings containing, respectively, nitrogen, oxygen, and sulfur atoms—are prepared by nucleophilic displacement reactions such as those used to prepare the corresponding three-membered rings, as shown:

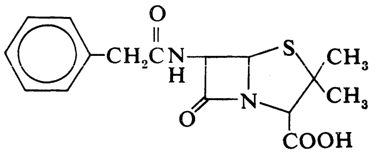


azetidine: Z=NH
oxetane: Z=O
thietane: Z=S

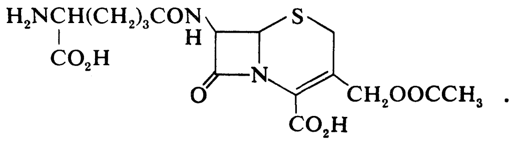
With the four-membered rings, however, the reactions proceed less readily than do the analogous reactions for the formation of the three-membered rings. Similarly, the ring-opening reactions of the four-membered heterocycles qualitatively resemble those of the corresponding three-membered rings, but they occur rather less readily. The most important heterocycles with four-membered rings are two related series of antibiotics, the penicillins and cephalosporins, both of which contain the azetidinone ring:



The chemistry of azetidinones, or β-lactams, as they are also called, was explored thoroughly during the intensive research into penicillin structure and synthesis that took place during World War II, a practical synthesis of penicillin not being achieved, however, until 1959. The complete chemical structures of a representative penicillin and a cephalosporin are shown below:

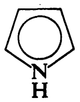


benzylpenicillin

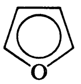


cephalosporin C

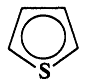
Five-membered rings with one hetero-atom. The parent aromatic compounds of the family—pyrrole, furan, and thiophene—have the structures shown:



pyrrole

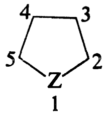


furan

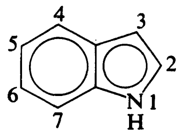


thiophene

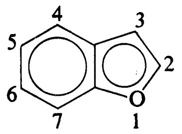
Five-membered heterocycles with one hetero-atom conform to the general structure:



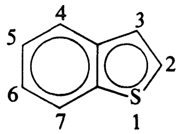
with numbering to indicate substituents as shown. The saturated analogues are called pyrrolidine, tetrahydrofuran, and thiophane, respectively. The bicyclic compounds containing a benzene ring fused to a pyrrole, furan, or thiophene ring are known, respectively, as indole, benzofuran, and thionaphthene; their structures are as shown:



indole



benzofuran



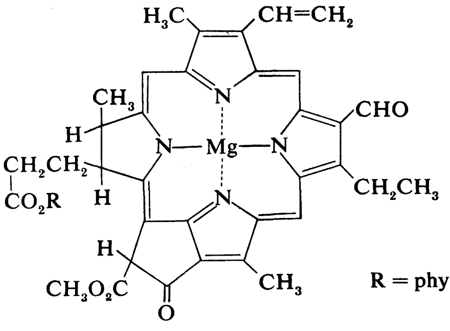
thionaphthene

The common numbering system is shown only for indole.

As mentioned earlier, pyrrole occurs in bone oil, in which it is formed by the pyrolytic decomposition of proteins (induced by strong heating). Reduced pyrrole rings are found in the essential amino acids proline and hydroxyproline, which are components of most proteins.

Occurrence of pyrrole derivatives

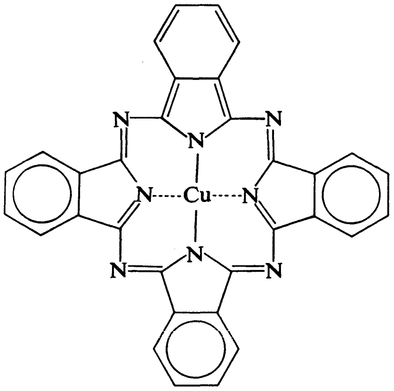
Pyrrole derivatives are of widespread natural occurrence: the blood pigment hemoglobin and related compounds, the chlorophylls, and also vitamin B_{12} all are formed from four pyrrole nuclei arranged in a large ring, such as that of chlorophyll b, shown below:



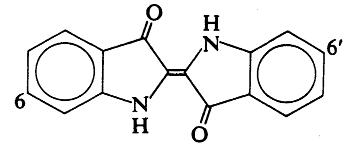
R = phytyl, $C_{20}H_{39}$

chlorophyll b
(plant pigment)

The phthalocyanins are a group of synthetic pigments that contain four benzopyrrole rings linked together in a large ring; a typical member of the family is Monastral blue with the structure shown:



The important vat dye indigo was formerly obtained from natural sources but is now synthesized on a large scale.

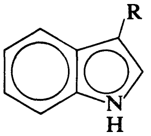


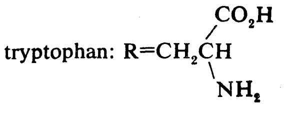
indigo

Tyrian purple, a dye used in classical times, is 6,6′-dibromoindigo (with bromine atoms at the numbered carbons).

Indole compounds

Other indole derivatives occur in nature. Tryptophan is an essential amino acid found in most proteins, and one of its degradation products is skatole, found in feces. Indole-3-acetic acid is a plant growth hormone, and many plant alkaloids contain indole rings. The structures of these compounds are:

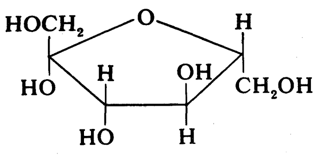




tryptophan: $R{=}CH_2CH(CO_2H)(NH_2)$

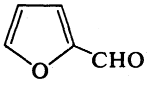
skatole: $R{=}CH_3$
indole-3-acetic acid: $R{=}CH_2CO_2H$

All carbohydrates, the biochemical family that includes the sugars and starches, are composed of one or more sugar (monosaccharide) units. These sugars are polyhydroxyaldehydes or -ketones (carbonyl compounds), which frequently exist as furanoside, that is, as a cyclized form with a five-membered, furanose, ring—*e.g.*, fructose, or fruit sugar—as the fructofuranose shown below:

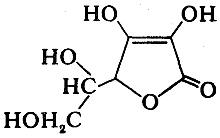


fructofuranose

Dehydration (removal of water) of certain carbohydrates yields furan derivatives; of great commercial importance is the conversion of a carbohydrate in corncobs into furan-2-aldehyde, or furfural, which is extensively used as a solvent, in the manufacture of plastics, and in the preparation of other furan derivatives. Many other reduced furans occur naturally, including vitamin C. The structures of furfural and vitamin C are as follows:

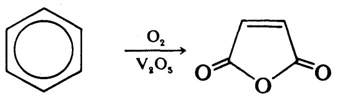


furfural

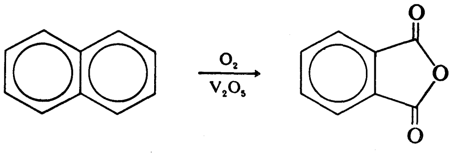


ascorbic acid
(vitamin C)

Other derivatives of furan of industrial importance are the solvent tetrahydrofuran prepared from furfural, and maleic and phthalic anhydrides, important constituents of resins and plastics. These compounds are prepared in quantity by the oxidation of benzene and naphthalene, respectively, as shown:

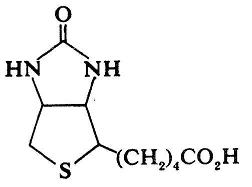


benzene maleic anhydride



naphthalene phthalic anhydride

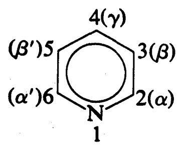
Thiophene and related compounds are found in coal tar and crude petroleum. The most important biologically occurring derivative is the vitamin biotin, with the structure:



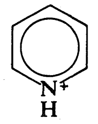
biotin
(vitamin H)

Pyridine
and
derivatives

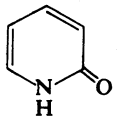
Six-membered rings with one hetero-atom. The nomenclature used for the various monocyclic nitrogencontaining six-membered ring compounds is as shown below, with numbering as shown for pyridine, arabic numerals being preferred to Greek letters (though both systems are used). The pyridones are aromatic compounds because of contributions to the resonance hybrid from charged canonical forms such as that shown for 4-pyridone. Mono-, di-, and trimethylpyridines are called picolines, lutidines, and collidines, respectively, with the position of the methyl groups denoted by the appropriate numbers; *e.g.*, 2, 4, 6-collidine. Pyridine-2-, -3-, and -4- carboxylic acids also have widely used trivial names; picolinic, nicotinic (derived from nicotine, of which it is an oxidation product), and isonicotinic acid, respectively. Pyridine and the picolines, lutidines, and collidines occur in coal tar and bone oil. Pyridine



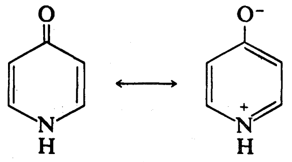
pyridine



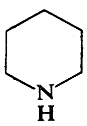
pyridinium ion



2-pyridone,
2(1H)-pyridone,
or 1,2-dihydro-
2-oxopyridine

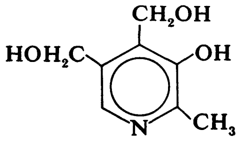


4-pyridone

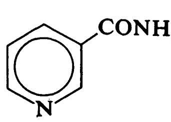


piperidine

derivatives are also of great biological importance: vitamin B$_6$ is pyridoxine, and another B-group vitamin is nicotinamide. Their structures are as follows:

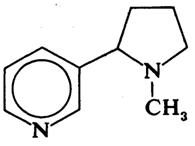


pyridoxine
(vitamin B$_6$)

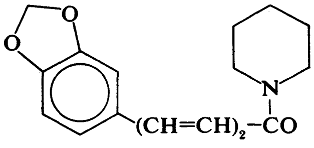


nicotinamide

Coenzymes I and II are derived from nicotinamide, and codecarboxylase from pyridoxine. Many alkaloids are derived from pyridine or piperidine, among them nicotine (in tobacco) and piperine (in pepper), with the structures shown.

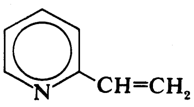


nicotine

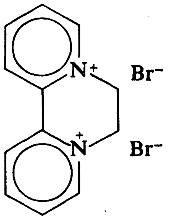


piperine

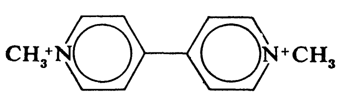
Pyridine, which is now prepared catalytically on a large scale from tetrahydrofurfuryl alcohol and ammonia, is an important solvent and intermediate used to make other compounds. Vinyl-pyridines, such as



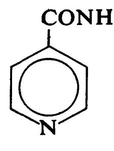
are important monomers used to make plastics. Pharmaceutically important pyridines include the tuberculostat isoniazid, the local anesthetic α-eucaine, the sulfa drug sulfapyridine, and the respiratory stimulant coramine. The chemical structures of these compounds are as follows:



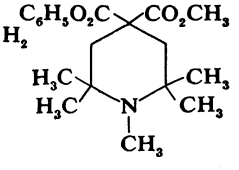
diquat



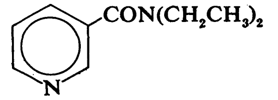
paraquat
(herbicide)



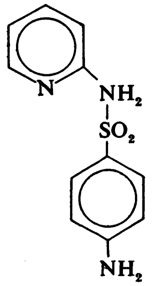
isoniazid



α-eucaine

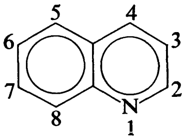


coramine

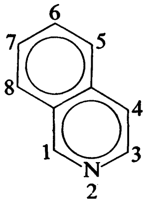


sulfapyridine

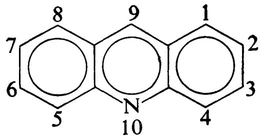
The structures of two isomeric benzopyridines and of two isomeric dibenzopyridines, with their common names and accepted numberings, are shown in the diagrams below:



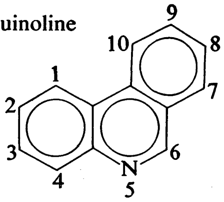
quinoline



isoquinoline



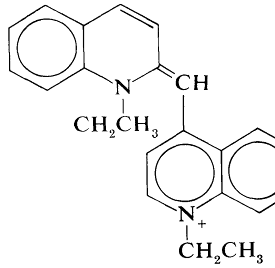
acridine



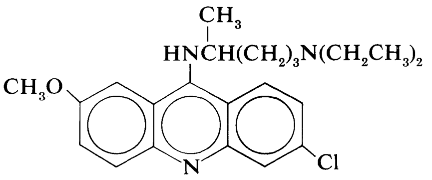
phenanthridine

Coal tar as a source

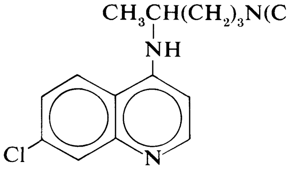
All four of these compounds and some of their alkyl derivatives have been obtained from coal tar. Each of them also is the parent substance of a class of alkaloids; of these, the quinoline (*e.g.,* quinine) and the isoquinoline (*e.g.,* morphine) groups are particularly well known. (For detailed coverage of this family of compounds, see the section *Alkaloids* below.) Important synthetic derivatives of the benzo- and dibenzopyridines include members of the family of cyanine dyes, some of which—*e.g.,* ethyl red—are photographic sensitizing agents and others are useful antiseptics for staphylococcal organisms; antimalarials such as mepacrine (also called quinacrine or Atabrine) and chloroquine; antibacterials such as acriflavine (Trypaflavine); trypanocides such as phenidium chloride; and the reagent oxine (8-hydroxyquinoline or 8-quinolinol), used in analytical chemistry. The structures of these compounds are as follows:



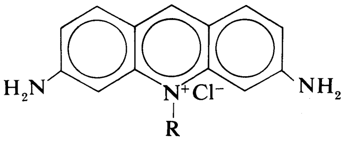
ethyl red



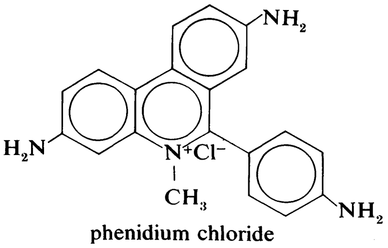
mepacrine



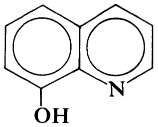
chloroquine



acriflavine (R = H or CH_3)

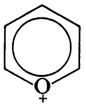


phenidium chloride

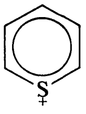


oxine

The parent six-membered, aromatic, monocyclic, oxygen and sulfur compounds are the pyrylium and thiopyrylium compounds bearing positive charges (cations) with the following structures:

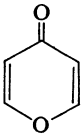


pyrylium cation

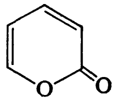


thiopyrylium cation

An uncharged aromatic (completely conjugated) six-membered ring containing an oxygen or sulfur atom is possible only if the ring contains a carbonyl group as in the pyrones:

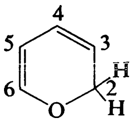


4-pyrone



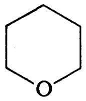
2-pyrone

The pyrans contain extra hydrogen atoms, the position of which is indicated by a number followed by an *H,* as shown:



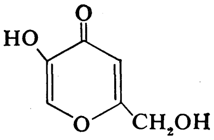
2[*H*]-pyran

Certain sugars are called pyranoses because they contain six-membered tetrahydropyran rings, with the structure:

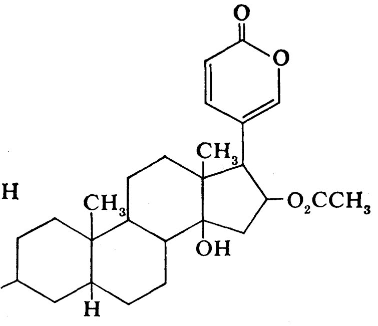


tetrahydropyran

Pyrone derivatives are present in natural products; *e.g.,* kojic acid, an antibiotic derived by the action of certain molds on starches or sugars, and bufotoxin—a poisonous ester of the steroid bufotalin—obtainable from the skin glands of toads. The structures of kojic acid and bufotalin are shown below:

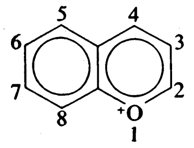


kojic acid

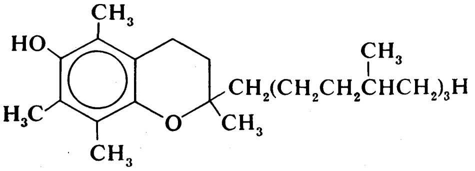


bufotalin

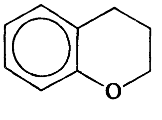
The benzopyrylium cation is the parent of a large number of natural products. Vitamin E is a substituted chroman, whereas dicoumarol, a blood anticoagulant, is a derivative of coumarin. The structures of these compounds are given below.



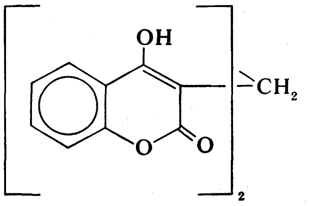
benzopyrylium ion



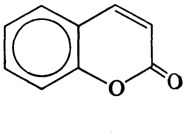
vitamin E



chroman

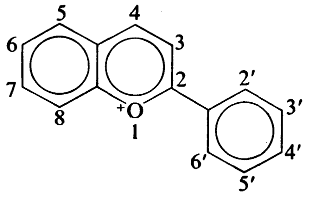


dicoumarol



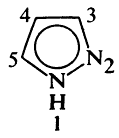
coumarin

The flavylium cation is the parent of the anthocyanidines, substances that are combined with sugars to form the anthocyanin pigments, the common red and blue colouring matters of flowers and fruits. The flavylium ion has the following structure:

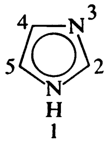


flavylium ion

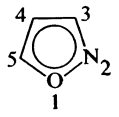
Five- and six-membered rings with two or more hetero-atoms. The names and numbering systems for the five-membered hetero-aromatic rings with two hetero-atoms are shown below:



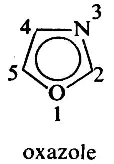
pyrazole



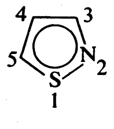
imidazole



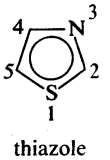
isoxazole



oxazole

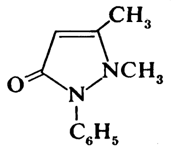


isothiazole

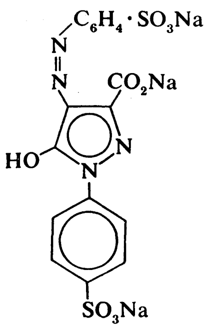


thiazole

Few pyrazoles occur naturally; the antipyretic (fever-reducing) and analgesic (pain-killing) antipyrine and the food and photographic dye tartrazine are important synthetic pyrazoles, with the following structures:



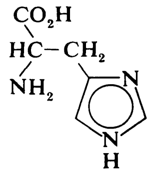
antipyrine



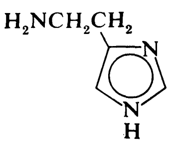
tartrazine

Imidazoles

Imidazoles are most important biologically; histidine, for example, is an essential amino acid, known to be of particular importance in enzyme reactions. A breakdown product of histidine, called histamine, is thought to be responsible for the development of allergies, hence the importance of antihistamine drugs. Histidine and histamine have the structures:

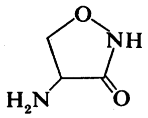


histidine



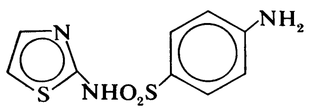
histamine

The antibiotic cycloserine, with structure shown, is one of the few naturally occurring isoxazoles:



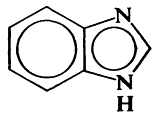
cycloserine

Thiazoles are of great biological importance. This ring system occurs in thiamine (vitamin B_1) and penicillin (see above). Sulfathiazole, an important sulfa drug, has the structure:



sulfathiazole

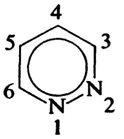
Most bicyclic systems derived from these five-membered rings are named systematically; *i.e.*, by use of the prefix "benz-" to indicate the presence of the benzene ring. Benzimidazole, for example, is the name for the compound:



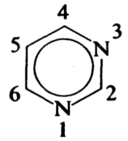
benzimidazole

A benzimidazole unit occurs in vitamin B_{12}.

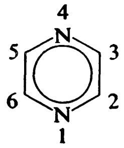
The three monocyclic diazines (six-membered-ring compounds with two nitrogen atoms) are named and numbered as follows:



pyridazine

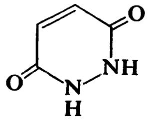


pyrimidine

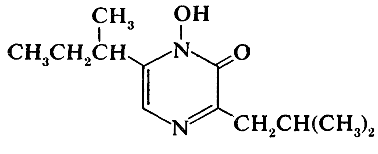


pyrazine

The pyridazine derivative maleic hydrazide is a herbicide, and some pyrazines occur naturally; *e.g.*, the antibiotic aspergillic acid. The structures of these compounds are:

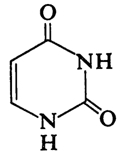


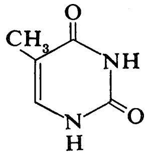
maleic hydrazide

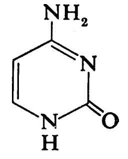


aspergillic acid

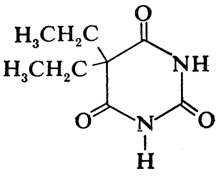
The pyrimidines, however, are the compounds of this group that are most important. Uracil, thymine, and cytosine, for example, with the structures shown, are constituents of nucleic acids.

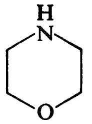

uracil


thymine

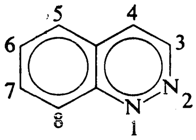

cytosine

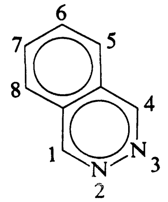
Vitamin B_1 contains a pyrimidine ring, and synthetic barbiturates, such as Veronal, with the structure given here, are widely used drugs. Various oxazines and thiazines are known but are of little importance, except for the solvent morpholine:

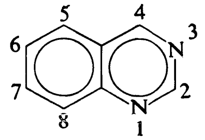

veronal

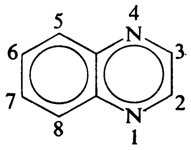

morpholine

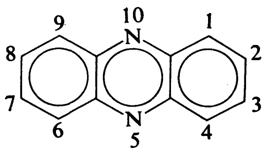
Many of the benzodiazines have common names; the structures and numbering of these compounds are as follows:


cinnoline

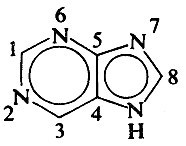

phthalazine

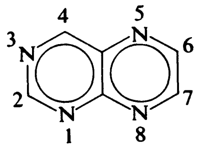

quinazoline

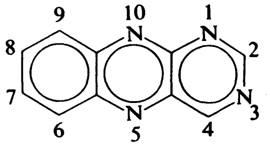

quinoxaline

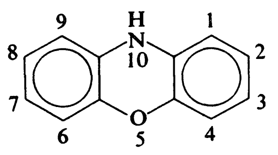

phenazine

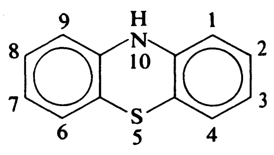
Five other polycyclic systems in this general family are of some significance; their structures and numbering system also are given:


purine

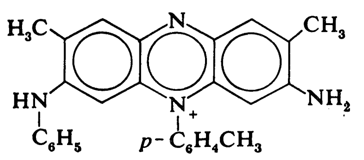

pteridine


alloxazine


phenoxazine

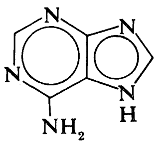

phenothiazine

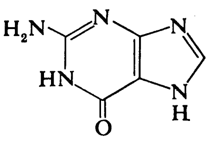
A few quinoxaline alkaloids exist; there are some phenazine natural products, and phenazine dyes are used for fabrics, the first synthetic dyestuff, Mauve, being historically important—its structure is as follows:


mauve

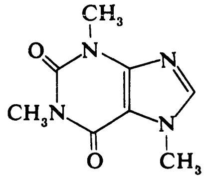
Purines and pteridines

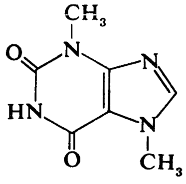
Biologically, however, the purines and pteridines are of more importance. Two purines, adenine and guanine, occur, together with the pyrimidine bases mentioned previously, in all nucleic acids.

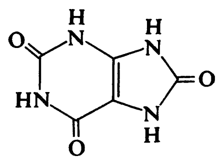

adenine


guanine

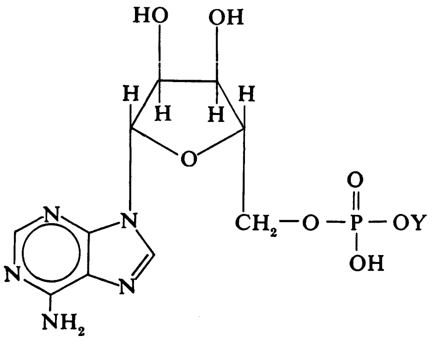
Other natural purines include caffeine (found in tea and coffee), theobromine (found in cocoa), and uric acid (a metabolic product). The structures of these compounds are as follows:


caffeine


theobromine

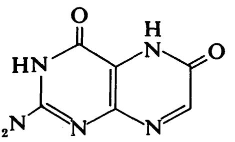

uric acid

Adenosine mono-, di-, and tri-phosphate (AMP, ADP, ATP, respectively) with the structures shown, are important in biological energy processes.

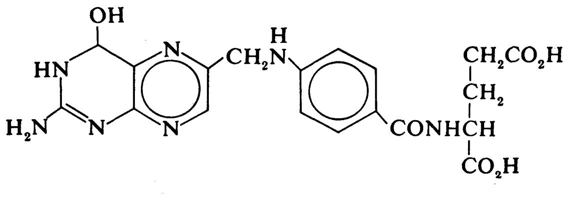


adenosine monophosphate: $Y = H$
adenosine diphosphate: $Y = PO_3H_2$
adenosine triphosphate: $Y = P_2O_6H_3$

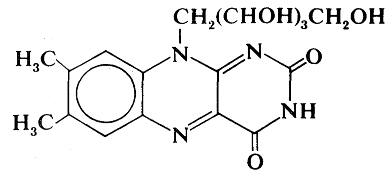
The biological significance of pteridine compounds has become apparent. The first pteridines were discovered as pigments of butterfly wings; *e.g.*, one with the structure:


2-amino-4,6-pteridinedione

Folic acid, also a pteridine, is an important growth factor.

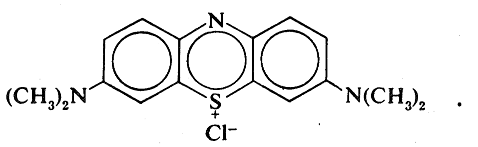

folic acid

Riboflavin, a derivative of alloxazine, is a B-vitamin:



riboflavin

Phenothiazine is an anthelmintic (worm-killing agent) as well as the parent of a family of dyes, such as methylene blue, a substance widely used as a biological stain and as an oxidation-reduction indicator. The structure of methylene blue is:



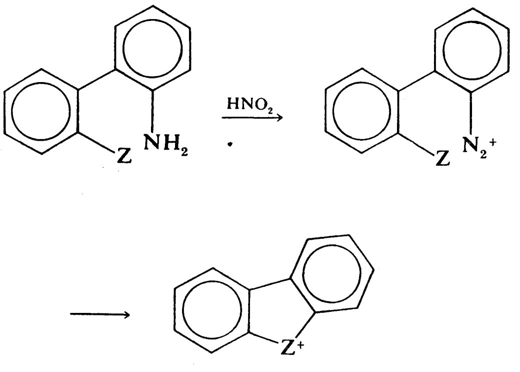
methylene blue

High stability

Rings with seven or more members. As the size of the ring increases, the range of compounds obtainable by variation of the number, type, and location of the heteroatoms increases enormously. Nevertheless, the chemistry of heterocyclic compounds with seven-membered rings, or larger, is still relatively undeveloped, chiefly because the formation of such rings is generally difficult, just as is the formation of large carbocyclic rings. These larger rings, however, can be very stable, once they are formed, and knowledge of their chemistry is likely to increase rapidly in the near future. Some larger ring heterocycles exist in natural products; the porphyrins are widely distributed as pigments; *e.g.*, the chlorophylls and hemoglobins mentioned above.

Rings with uncommon hetero-atoms. Although nitrogen, oxygen, and sulfur are the common atoms found together with carbon in heterocyclic rings, a large number of other elements also form such rings—of greater or lesser stability. Such compounds are as yet of little practical importance. The principal ones known to date contain elements of the classes described below.

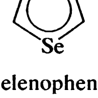
Halogens. Certain cyclic chloronium, bromonium, and iodonium ions have been prepared by the following reaction:



Z = Cl, Br, or I

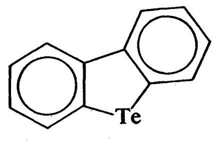
Of these, only the iodine derivative has much stability.

Selenium and tellurium. Many selenium heterocycles are known. Selenium shows much similarity in its behaviour to sulfur; hence selenophene, with the structure shown,



selenophene

resembles thiophene quite closely. Tellurium heterocycles are more rare, and less stable; an example is dibenzotellurophene:

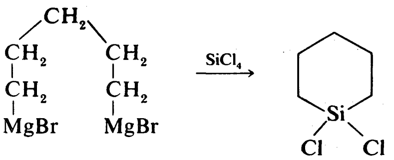


dibenzotellurophene

Phosphorus, arsenic, and antimony. Phosphorus, arsenic, and antimony, all members of Group V of the periodic table of elements, form a closely related group of heterocycles. There is, however, little similarity between their properties and those of the corresponding derivatives of nitrogen, a member of the same group. With few exceptions, no hetero-aromatic compounds are formed by these elements, and the cyclic derivatives resemble their open-chain analogues.

Silicon, germanium, and tin. Although silicon, germanium, and tin fall in the same periodic group as carbon, their atoms form neither stable chains nor double bonds; hence no hetero-aromatic derivatives of these compounds are known. Many saturated heterocyclic derivatives are known, however, including cyclic silanes, which may be prepared as shown by the following example:

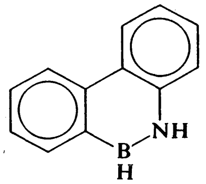
Silanes



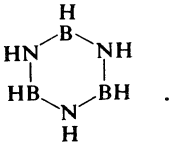
1,1-dichlorosilacyclohexane

All these elements, particularly silicon, readily form rings in which the atoms of these elements alternate with oxygen atoms. Considering this fact, many silicates are, strictly speaking, heterocyclic compounds, but their chemistry is always considered with that of other inorganic compounds, which they closely resemble.

Boron. Although boron heterocyclic chemistry is not yet extensively developed, it has a large potential. A boron and a nitrogen atom together contain the same number of electrons as two carbon atoms. Not surprisingly, it has been found that a boron-nitrogen unit can replace a carbon-carbon unit in benzenoid compounds to give stable hetero-aromatics; a good example is 9-aza-10-boraphenanthrene; its structure is as follows:



Borazine, or borazole, which has been referred to as inorganic benzene, has the following structure:



(A.R.K.)

Dyestuffs and pigments

Dyestuffs and pigments are intensely coloured substances used for imparting colour to other materials. Technically, dyestuffs are soluble in the medium in which they are applied, whereas pigments are insoluble.

Colour in materials originates either from organic or inorganic substances present in the material, or from a purely physical effect, such as that observed in oil films on water. Organic colouring matters are compounds of the element carbon, including animal or vegetable colouring matters, as well as synthetic materials, like the coal-tar dyes. Colour in inorganic substances is commonly observed in rocks, Organic and inorganic colouring matters

minerals, ceramics, fired enamels, coloured glass, precious stones, and certain pigments. Colours produced by physical effects result from dispersion, refraction, or scattering of light. Examples of such physically produced colours are the rainbow; the blue of the sky; the iridescence of opal; and the blue, green, and violet colorations of birds, butterflies, fish, and insects. In some living organisms organic and physical coloration occur together.

There is scarcely any manufactured product that does not at some stage incorporate or require the use of products of the dyestuff and pigment industry. About 7,000 different dyestuffs are currently made and sold under some 35,000 different trade names, and about 200 new ones are introduced every year.

Reasons for the existence of the great number of commercial dyestuffs include the variety of fibres and other materials requiring coloration and the fact that colour in textiles must withstand a variety of stresses, depending upon the methods of manufacture and use.

Use of pigments

All dyestuffs and many pigments are complex organic compounds, mainly prepared synthetically from chemical products of the coal-tar and petrochemical industries. The syntheses consist typically of sequences of carefully regulated procedures carried out under strict scientific control, and they are the result of extensive research. Dyes are ordinarily dissolved in water for application to textiles, paper, leather, and many other substances. Pigments are used as finely ground solid particles rather than as solutions. They are employed, for example, in paints, printing inks, and plastics. Some pigments are entirely inorganic in origin; *e.g.,* iron oxide. These are cheap, but rather dull in hue. The brighter pigments are organic compounds, and in fact may simply be dyestuffs insoluble in the medium. Many products are thus used in both capacities; *i.e.,* in soluble form for dyeing textiles, paper, and other materials and in insoluble form for colouring such substances as paints and inks.

All organic coloured substances, whether natural animal and vegetable materials or synthetic dyestuffs, rely for their colour on certain aspects of their chemical structures. Organic dyestuffs and pigments, with which this section largely deals, are classified as colorants.

HISTORY

Development of natural dyestuffs. The Bible tells us that Moses was instructed to accept gifts of "blue, and purple, and scarlet and fine linen and goats' hair." These names of colours most probably referred to materials dyed with dyes known as indigo, Tyrian purple, and kermes. Indigo was extracted from plants of the *Indigofera* genus, Tyrian purple from mollusks of the genera *Murex* and *Purpura,* and kermes from a dried insect, *Kermes ilicis.* The Bible, therefore, illustrates the antiquity of the craft of dyeing and the variety of natural products used in ancient times as sources of dyestuffs. Dyed fabrics discovered in Egyptian tombs show that dyeing was practiced at least as early as the 25th century BC. It is believed that the craft originated in India (the name indigo is derived from *indikos,* the Greek word for Indian) and spread westward to Persia, Phoenicia, and Egypt.

The method of producing a variety of hues by mixing red, blue, and yellow dyestuffs was known in ancient times. Fabrics found in caves in the Dead Sea area and dating from AD 135 were found to have been dyed with various mixtures of safflower yellow on alum mordant (fixative), kermes on alum, and indigo blue.

Until the 19th century all dyestuffs were natural products, obtained mainly from vegetable sources. Many different plants were used. The craft of dyeing demanded a high degree of skill, and details of its recipes were largely secret. Some of the most important natural dyestuffs that were in use until the advent of synthetic products are shown in Table 41.

Table 41: Sources of Some Important Natural Dyestuffs

Yellow	
Persian berries*	the fruit of a *Rhamnus* species
Weld*	the leaves, seeds, and stem of the plant *Reseda luteola*
Fustic*	the wood of *Morus* or *Maclura tinctoria* (Central or South America)
Quercitron bark*	from *Quercus nigra* or *tinctoria* (U.S.)
Brown	
Catechu or cutch*	acacia, etc., wood (Far East)
Reds	
Kermes	a dried insect, *Kermes ilicis* (Mediterranean region)
Lac dye	the shellac-producing insect *Coccus lacca* (Far East)
Cochineal*	the dried female of the insect *Coccus cacti* (Mexico, etc.)
Archil and cudbear	prepared by fermentation of various lichens, *e.g., Roccella tinctoria*
Madder (Turkey red)*	the root of *Rubia tinctorum* (Asia Minor)
Safflower	the dried flower head of the Dyer's Thistle (*Carthamus tinctorius*)
Blue	
Indigo	prepared from the leaves of various *Indigofera* plants, also the *Isatis tinctoria* or woad plant
Violet	
Tyrian purple	prepared from various mollusks; *e.g., Murex brandaris*
Black	
Logwood*	the wood of the West Indian tree *Haematoxylon campechianum*†

*Used with mordants. †Before the discovery of America, blacks were prepared from tannins (*e.g.,* from oak galls or sumac) and salts of iron.

With few exceptions these natural dyestuffs have no affinity for cellulose, and thus they could be applied to cotton only with the aid of an auxiliary inorganic compound or mordant (Lat. *mordere,* "to bite"), which precipitated the colouring matter in a less soluble form in the fibre. Mordants were also used on wool and silk, both to improve fastness and to obtain a wide variety of hues, different mordants yielding different hues with the same dye. The mordants in common use in ancient times were alum (potassium aluminum sulfate) and salts of iron, copper, or tin. Salts of chromium came into use about 1850. Eventually, on account of the superiority in fastness of the dyeings they produce, chromium mordants almost entirely displaced all other metal salts. Chromium mordants are still widely used for dyeing wool and to a lesser extent for silk and nylon, but are no longer of importance for cellulose fibres.

Ancient use of mordants

Following the introduction of synthetic dyes in the 19th century, natural dyestuffs have steadily declined in importance. Now only one, logwood, employed to dye nylon black, is used in any significant quantity.

Development of the chemistry of dyestuffs and dyeing. "The Liquors that Dyers imploy to tinge," wrote Robert Boyle, an English physicist, in 1664 in the first book dealing with the science of coloration, "are qualified to do so by multitudes of little Corpuscles of Colour . . . insinuating themselves into, and filling all the Pores of the Body to be Dyed, ..." Even today this remains a good description of the chemistry of dyeing.

As early as the mid-17th century, therefore, leading scientists were concerned with the processes of dyeing, and one of the first actions of the Royal Society after its foundation was to assist dyers by publication of a paper, "An Apparatus to the History of the Common Practices of Dyeing," read to them by Sir William Petty in 1662, describing the methods of dyeing then in use.

Petty, in his paper, classified dyestuffs into the fundamental colours, red, yellow, and "blew," and discussed the use of alum as a mordant, suggesting that it first is taken up by the fibre, where it then becomes combined with the dye. Thus, the dye is fixed in the fibre. This is also the modern view.

Early chemical interest in dyeing

From that period onward the craft of dyeing was aided by the discoveries of chemists and, especially in the period 1670–1830, by a series of French chemists, whose interest appears to have stemmed from the action of Louis XIV's comptroller general of France, who, in order to establish high standards in the wool textile industry, laid down regulations controlling dyeing methods. Distinction was made at that time between *bon teint* and *petit teint,* or fast and nonfast colours; *i.e.,* dyeings that were fast or fugitive, respectively, to light and washing. When in 1729 the regulations were redrafted, the first system of controlled testing of fastness of dyed textiles was introduced.

During the early part of the 19th century, dyeing continued to attract some scientific notice. Even greater interest was created by the advent of synthetic dyestuffs. William Henry Perkin's discovery (1856) of mauve, the first commercial synthetic dyestuff (see below), and the subsequent development of a great variety of dyestuffs of known chemical constitution initiated the modern expansion of the subject into a distinct branch of chemistry. This growth has been particularly noteworthy since World War I. In addition to the availability of a continually growing number of dyestuffs, the reasons for this development include improved instrumentation, increased knowledge of the structure of fibres, and the application of classical chemical principles to the dyeing process.

The introduction of synthetic dyestuffs challenged organic chemists with the problem of determining the molecular structure of these complex materials and opened the way to the development of theories of the causes of colour in organic compounds in general. The idea of valence (fixed numbers of bonds for atoms of different kinds), and in particular the quadrivalence of carbon, had already been formulated when the new dyestuffs began to appear. Shortly afterward the ring structure of the benzene molecule and the three-dimensional nature of organic molecules in general were suggested. These advances greatly assisted in the understanding of the molecular structures of organic dyestuffs.

Chromophores and auxochromes

One of the most important contributions to the understanding of the relationship between colour and structure was the so-called "chromophore" theory proposed in 1876 by a German chemist, Otto N. Witt. In every dye molecule, Witt identified a chromophoric ("colour-bearing") group of atoms together with one or more "auxochromes"; *i.e.,* associated groups whose function was to intensify the colour. In 1888 an English chemist, Henry Edward Armstrong, showed that many chromophores could be depicted as quinones (oxidized structures derived from the benzene ring). These theories are substantially those held today, although they have been broadened by recognition that the function of the chromophore is to produce a strong absorption of radiation, either in the ultraviolet or in the visible region of the spectrum, and that of the auxochrome is to shift the absorption to give more intense coloration (see below *Dependence of properties upon molecular structure*).

The development of the quantum theory in the 20th century also brought greater insight into the mode of interaction between light and organic molecules. In current terminology, the absorption of visible radiation by a dye raises the electronic energy of the molecule to a so-called "excited state." The energy must then be dissipated either as heat, as phosphorescence or fluorescence, or in a chemical reaction with the surroundings, in which case the dye decomposes and loses its original colour.

Development of synthetic dyestuffs. In 1856 William Henry Perkin, a chemistry student in London, discovered a purple colouring matter that had dyeing properties in a product he obtained by chemical treatment of aniline, a material from coal tar. This chance discovery, made in a simple laboratory Perkin had fitted out at his home, laid the foundation of the manufacture of synthetic dyestuffs and pigments, and indeed of the entire world-wide synthetic organic chemical industry.

The discovery of mauve

Perkin patented his dye and, with foresight and courage, in 1857, established a factory to manufacture what he at at first called Aniline Purple, or Tyrian Purple, and later Mauve, a French word for the mallow, a plant with a purplish flower. Perkin, moreover, developed new methods of applying the dyestuff and visited dyers to show them how they should be used.

Mauve was an immediate success, and the general method of its preparation, the oxidation of aniline, was varied and elaborated by other investigators to produce many new dyestuffs, first the crimson Fuchsine or magenta, and later, blues, violets, and green. These were the first basic or cationic (positively charged) dyestuffs, a class that later became overshadowed in importance by others, though in recent years interest in it has revived. Because of their origin, these earliest synthetic dyestuffs were called aniline dyes, a term still used. Unfortunately, their somewhat fugitive nature caused all synthetic dyestuffs to be considered inferior to the natural products, though this is certainly not the case.

In 1868 a team of German chemists proved that alizarin, the colouring principle of madder root, is a derivative of the chemical compound anthraquinone. (Figure 10 shows the colours yellow to purple obtained with pigments manufactured by modifying the molecular structures of anthraquinone compounds.) Their synthesis of alizarin from anthraquinone was the first preparation of a natural vegetable dyestuff by purely chemical means. Upon the introduction of synthetic alizarin, the use of the natural madder root died out. But, at the same time, new and related products were introduced. The technique of sulfonation (that is, the introduction of sulfonic acid groups), developed earlier, was applied from 1893 onward to the development of a new class of valuable bright, fast dyestuffs for wool, the sulfonated anthraquinone derivatives. This class is one of the most important for wool, and the unsulfonated counterparts are used to dye modern man-made fibres.

From *Journal of the Oil and Colour Chemists Association* (January 1963)

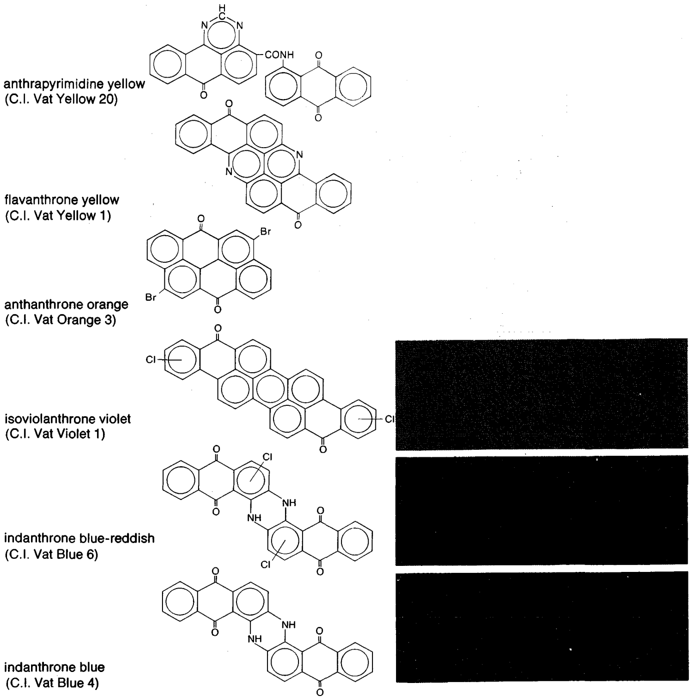


Figure 10: *Colours obtained from various anthraquinone pigments.*
By modifying the benzene rings of the anthraquinone group, a range from yellow to purple can be produced. A great many other derivatives of anthraquinone are used for industrial purposes.

The azo dyestuffs

Meanwhile, an entirely new type of colouring matter, the azo dyes, which now comprise the largest group of dyestuffs, was being developed. Johann Peter Griess, a young German, then working in the laboratory of an English brewery, discovered the diazonium compounds in 1858. These are very reactive substances, formed from the reaction of aromatic amines (nitrogen-containing organic compounds) with nitrous acid. Diazonium salts readily combine with aromatic amines or phenols to give deeply coloured substances containing the azo group (two nitrogen atoms joined by a double bond). The discovery was not, however, fully exploited until the 1870s, when the full tide of development of the azo dyestuffs set in, stimulated by the discovery that the addition of sulfonic acid groups gives highly water-soluble materials with excellent affinity for wool. By treating naphthalene derivatives with sulfuric acid, various sulfonated aminonaphthalenes and hydroxynaphthalenes were made, from which azo dyestuffs could be prepared in a range of shades—from orange to black. And by use of derivatives of pyrazolone instead of naphthalene, yellow azo dyestuffs also were prepared. In 1884 Congo red was produced; this is a bisazo dyestuff; *i.e.,* one containing two azo groups in its molecules. Congo red was the forerunner of a long series of successful dyestuffs, still of importance, named "direct cotton" or "direct" dyestuffs, which can dye cotton without a mordant.

Vat dyestuffs

Indigo for a time was the sole example of a "vat" dyestuff; *i.e.,* one applied from a "vat" containing alkali and a reducing agent in which the insoluble dyestuff was temporarily solubilized during dyeing. In 1901 Indanthrone blue, the first of the modern type of vat dyestuff, based on anthraquinone, was introduced; and this class, including dyestuffs of the very greatest fastness, has become of great importance.

Four other important new classes of dyestuffs were introduced in the 20th century:

1. The azoic dyestuffs. As early as 1880, a substance called β-naphthol had been applied in alkaline solution to cotton and then coupled with a diazonium compound to produce an insoluble dyestuff in the cotton fibre. Little progress was made with this type of dyeing, however, until Naphthol AS (a more complex chemical compound derived from β-naphthol) was discovered in 1912 and used to replace β-naphthol. Later a whole series of such compounds—and new bases to be diazotized and coupled with them—appeared, giving bright, fast dyes, ranging in colour from yellow to black, but especially useful for oranges and reds.

2. The disperse or nonionic dyestuffs. These substances were discovered in 1923 and at first were used with the then new cellulose acetate fibre, but later were also used for nylon and polyester fibres.

3. The phthalocyanine dyestuffs and pigments. Discovered in 1907 and 1927 by European chemists who failed to realize their importance, phthalocyanine derivatives (a class of metal-containing organic compounds) were rediscovered by accident in Scotland in 1928. The first member of this class, a brilliant blue, was found to have a remarkable molecular structure, resembling those of chlorophyll and hemoglobin, natural colouring matters. It proved to be a pigment of outstanding properties. Chlorinated and brominated derivatives give bright green dyes that together with the parent compound are used in very large quantities mainly as pigments; modified products are used for dyeing. Although no other useful shades have been obtained from phthalocyanine derivatives, quinacridone (a red compound discovered in 1955) is capable of preparation in red or violet forms as a pigment of fastness comparable to phthalocyanine.

The first reactive dyestuffs

4. The reactive dyestuffs for cellulose. First introduced in 1956, these dyes represent an entirely new principle, in that the dyestuff becomes a part of the actual fibre molecule and, therefore, is highly resistant to washing. Unlike other fast dyestuffs for cellulose, many reactive dyestuffs have very bright shades. Later, reactive dyestuffs adapted for wool and nylon were introduced.

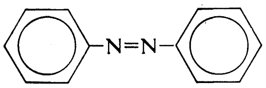
DEPENDENCE OF PROPERTIES UPON MOLECULAR STRUCTURE

To be useful, dyestuffs and pigments must have: (1) intense colour and (2) fastness; *i.e.,* resistance to the various chemical and mechanical stresses they meet in the manufacture or use of the finished coloured product. In addition, dyestuffs, but not pigments, must have: (3) solubility in the medium of application, which in almost all cases is water, and (4) ability to be adsorbed and retained by the fibre (substantivity) or to be chemically combined with it (reactivity). Pigments also have to meet unique requirements somewhat different from those of dyestuffs, including such physical and mechanical properties as the ability to flow in the media of application and insolubility in oils. These properties are favoured by making the molecule large, and insolubility is also promoted by particular dispositions of the substituent groups—that is, the peripheral units attached to the main body of the molecule.

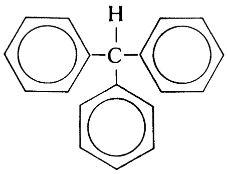
In all cases these properties depend in large part upon the molecular structures of the colorants themselves. The relationship between the properties of the dyestuffs and pigments and their molecular structure is dealt with in the following sections.

Colour. Organic colouring matters are complex unsaturated compounds—that is, compounds with multiple bonds between the atoms—and each such compound has a molecular structure containing one of the fundamental chromophores. Three of the most important chromophores in dyestuffs are called azobenzene, triphenylmethane, and anthraquinone. Their molecular structures are as follows:

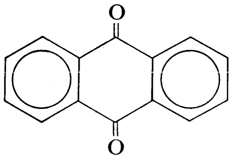
Structures of chromophores



azobenzene



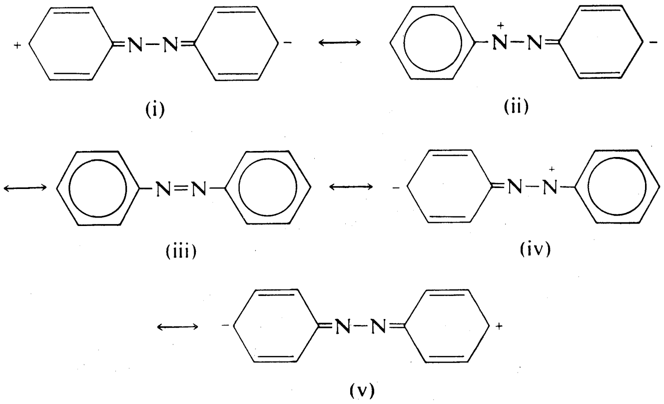
triphenylmethane



anthraquinone

in which the letters N, C, H, and O stand for atoms of nitrogen, carbon, hydrogen, and oxygen, respectively; the short single and double lines represent single and double bonds between them, and the hexagons with circles in them represent benzene rings, each composed of six carbon atoms. The chromophoric compounds themselves usually have very weak colours because the conjugated chain (the system of alternate single and double bonds, including the benzene ring) in the molecules is relatively short. If this chain is greatly extended, as in the natural carotene pigments, for example (see below), deep colour develops. Such extended chains, however, are not conveniently introduced into synthetic dyestuffs, which instead rely upon the above-mentioned substituent groups called auxochromes for colour intensification. Auxochromes generally are polar substituents—that is, atoms or groups favouring separation of electrical charges—including such substituents as chloro, nitro, hydroxy, methoxy, amino, methylamino, dimethylamino, and trifluoromethyl groups.

In the conjugated systems of alternate double and single bonds of the chromophores, some of the electrons, called pi electrons, are able to move through the molecule. It is possible to write various electronic structures of the molecule; *i.e.,* the so-called resonance forms. None of these structures describes the molecule as it really exists, but the molecule may be considered as a hybrid to which the imaginary structures contribute. Thus forms i–v below represent some of the forms contributing to the hybrid structure of azobenzene. Of these, iii is the most stable, because no energy is required to separate charges (as in the other forms).

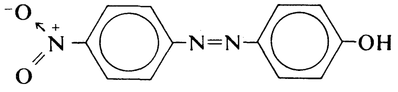


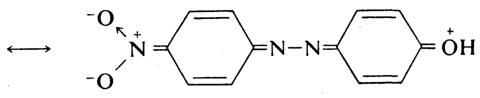
(i) (ii) (iii) (iv) (v)

In this representation, double-headed arrows are used between resonance forms of a single hybrid structure.

Influence of auxochromes

The function of the auxochromic groups is to increase the stability of the alternative resonance forms. They do so because they can retain electrical charges more readily than carbon atoms can. For example, when two auxochromic groups are attached to an azobenzene molecule, the resulting structure may consist of new hybrid forms, as in the example below:


(vi)


(vii)

In this instance the alternative electronic configuration vii is more stable than i, ii, iv, or v above, because the charges are now held on oxygen and nitrogen atoms, which can retain them more readily than carbon. The result is that this compound has a more intense colour than azobenzene does.

With the increase in stability of the contributing electronic structures of a molecule, its electrons may be considered to move more readily along the chromophore. Thus its natural frequency of vibration is decreased. A simple analogy can be drawn with a violin string: the longer the string, the lower is the frequency and the longer is the wavelength of the note it emits when plucked. In general, the wavelength increases with the length of the oscillator. This applies to the absorption as well as the emission of energy, whether it be sound energy or light energy, because any oscillator absorbs energy most readily at the wavelength of its natural frequency of vibration.

The chromophores described above absorb most energy at short wavelengths—that is, in the ultraviolet region of the spectrum; and consequently they appear colourless or have only a faint yellow colour. When an auxochrome is present the wavelength of absorption increases and moves into the visible region, thus causing colour to appear. For example, anthraquinone is almost colourless, but 1-methylaminoanthraquinone has an intense red colour. The resulting hue is said to be "deeper."

Table 42: Effect of Wavelength of Absorption on Colour Observed

region of maximum absorption*	colour seen†
Ultraviolet	none
Violet	yellow
Blue	orange
Green	bluish-red
Yellow	violet
Orange	blue
Red	blue-green

*Listed in order of increasing wavelength. †Listed in order of increasing depth of hue.

Light absorption and colour

Table 42 shows the changes that occur in the colour of a given chromophoric system with the progress of its main absorption band across the spectrum from short to long wavelengths. Such a succession of changes can often be produced by substitution of the chromophore with a succession of selected auxochromes, one or more at a time, if these auxochromes have increasing effectiveness in enhancing the mobility of the electrons of the chromophoric system. In the case of azo dyestuffs, for example, the hue can be deepened by substitution in either of the aromatic nuclei attached to the azo group, and if both are substituted, the effect is further enhanced.

The intensity of the colour of dyestuffs is among the highest found in any type of substance. The degree of this intensity, known as the molar extinction coefficient, is measured at the wavelength of maximum absorption of light. The molar extinction coefficient for dyestuffs is about 10^4. In practice, this means that a 1 gram percent solution of a dyestuff only 1/100th of a millimetre deep reduces the intensity of the most strongly absorbed light to about one-half of the original value. There is another way to visualize the colour intensity of dyestuffs; it requires from about 10 to 100 single layers of dyestuff molecules, placed one on top of another, to produce a visible colour; and, allowing for the minimum area the substance must have in order to be seen by the naked eye under the best conditions, a total of about 10,000 molecules, or about 10^{-18} gram, must be concentrated together before the colour becomes visible.

Fastness to light and to washing

Fastness. Ability to withstand washing without changing colour, "wash fastness," and ability to withstand exposure to light, "light fastness," are two of the properties of most importance to the user of coloured fabrics. Dyestuffs that are chemically bonded to the fibre—*i.e.,* reactive dyestuffs—or those that are present as water insoluble particles of low solubility in soap solutions, such as vat dyestuffs, have the highest wash fastness. Water-soluble dyestuffs often have lower wash fastness, but the resistance to removal by soap varies greatly from one dyestuff to another, fastness generally being aided by high molecular weight of the dyestuff molecule and a minimum number of solubilizing groups.

All organic colouring matters fade when exposed to light; hardly any inorganic colours, and no physically produced colours, do. Light fading is the most complex of the reactions that dyestuffs undergo during use, and a great deal of research has been devoted to discovering its causes, which are not yet fully understood.

The light fastness of any dyed material depends not only on the nature of the dyestuff itself, but also on many other factors, including the method of application, the depth of shade, the type of fibre and its moisture content, the relative humidity and oxygen content of the air, and the nature of the illumination.

Solubility. Because normal dyeing takes place from aqueous solutions, the dyestuff must contain substituent groups conferring solubility in water. Table 43 includes the most important solubilizing groups used in dyestuff molecules.

Solubility in water may be needed only temporarily during the actual dyeing operation, the dye later being treated so that it becomes water-insoluble on the fibre. Such temporary solubility may be obtained with an insoluble, nonionic dyestuff molecule containing quinone groups, for example, by chemical reduction (addition of electrons) in the presence of alkali to produce a salt of the corresponding hydroquinone; the original quinone is then regenerated on the fibre after dyeing by oxidation. This

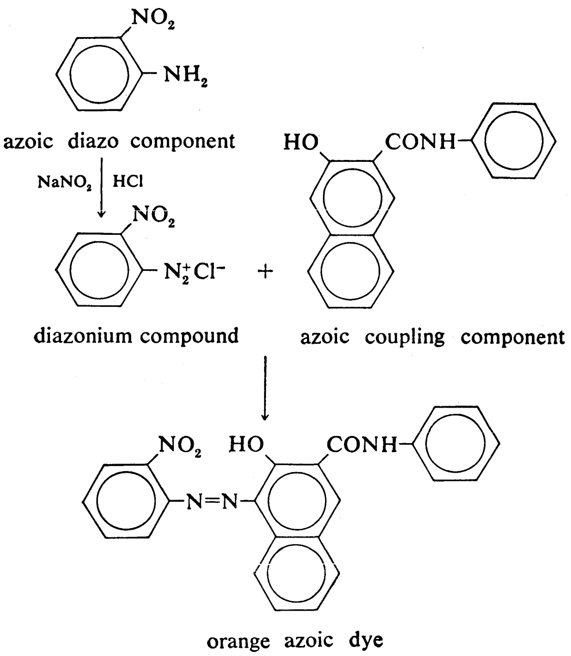
Table 43: Solubilizing Groups Used with Various Classes of Dyestuffs

name	structure*	type of dyestuff
Permanent		
Sodium sulfonate	$-SO_3Na$	direct cotton, acid wool, chrome mordant; 1:1 dye–metal complex for wool
Amine hydrochloride	$-\overset{+}{N}H_3Cl^-, -\overset{+}{N}R_3Cl^-$	basic, for cellulose, wool, silk, and acrylic fibres
Hydroxy, amino, and sulfamido groups	$-OH, -NH_2, -SO_2NH_2$	disperse, for cellulose acetate, nylon, polyester fibres; 2:1 dye–metal complexes for wool and nylon
Temporary		
Sodium phenolate	$-O^-Na^+$	certain classes of dyestuffs for cellulose; *i.e.,* naphthols for subsequent azo coupling on the fibre; and vat dyestuffs
Isothiouronium groups	$\left.-CH_2SC\begin{matrix}\diagup NR_2\\ \overset{+}{\diagdown} NR_2\end{matrix}\right\}Cl^-$	phthalocyanine dyestuffs for cellulose

*R = an alkyl or aryl group.

reduction-oxidation procedure is the basis of the vat dyeing process. Usually sodium dithionite (also called sodium hydrosulfite, formula $Na_2S_2O_4$) is the reducing agent used, and sodium hydroxide is the alkali, with the result that the sodium salt of the reduced form of the dye, known as the leuco compound, is formed. The entire process of alkaline reduction is known as vatting. The leuco compound is generally unstable in air, being rapidly oxidized back to the original dye.

In the method of dyeing cellulose fibres known as azoic coupling, a single soluble component, usually a naphthol, is applied to the fibre from solution in sodium hydroxide; this component is then converted into a water-insoluble azo dyestuff on the fibre by coupling with a soluble diazotized base. The procedure is illustrated in the following reaction sequence:



The so-called disperse dyestuffs used to colour cellulose acetates, nylon, and polyester fibres are employed as dispersions—that is, as fine suspensions of the solid in liquid—rather than as true solutions. These substances, however, are slightly soluble in water due to the presence of several nonionic water-attracting substituent groups on their molecules, and dyeing probably takes place from the very dilute solutions they produce.

There is considerable interest in processes for dyeing certain man-made fibres with disperse dyestuffs from solutions in organic solvents instead of water. Another process also being considered is the use of vaporized dyestuffs, at high temperature (above 200° C [392° F]).

Porosity and surface area

Attachment. Many solid materials adsorb dyestuffs from their solutions, but fibres do so particularly well because of their porosity and their ability to exert chemical forces upon the dyestuff molecules. There is an immense number of submicroscopic pores in a fibre, mainly directed parallel to the length of the fibre; there are, in fact, about 10 million in the cross section of a normal textile fibre. The total surface of the walls of these pores is extremely high, and in natural fibres, such as cotton or wool, the pore surface amounts to no less than about five acres per pound. This value is about one thousand times as great as the external surface of the fibres.

Since 1,000 to 10,000 dyestuff molecules placed one above another are needed to produce a deep coloration, a single layer of dyestuff molecules adsorbed on the walls of all the pores of a fibre would be more than enough to produce a satisfactory depth of shade. Actually, a large proportion of the pores are too small to allow a dyestuff molecule to enter, but, even so, enough dye is adsorbed to produce satisfactory depth of shade. For this action to occur, however, forces on the molecular scale must operate between the fibre molecules exposed in the pore surfaces, and the dyestuff molecules in the contiguous solution. The strongest of these forces are the forces of true chemical bonding by which the so-called reactive dyestuffs are fixed. Other types of dyestuff are attracted by a variety of weaker forces that operate between molecules. As already mentioned, some types of dyestuff are solubilized only during the dyeing operation; in such cases, after the subsequent insolubilization, the dyestuff particles remain held in the fibre by purely mechanical action.

With regard to the dyestuff molecule, attachments to the fibre resulting simply from intermolecular attraction are considered substantive, whereas those involving chemical bonding are considered reactive.

Substantivity. The presence of one or more of certain specific groups in a dyestuff molecule determines its substantivity, or affinity, for any given type of fibre, and hence its dyeing method. These groups include the following:

1. Anionic or cationic groups. These (respectively) negatively- or positively-charged groups confer solubility on dyestuffs and also affinity for the appropriate fibres with ionic properties.
2. Polar groups. Although these substituents are electrically neutral they do contribute to charge separation within a molecule. This effect increases the affinity of nonionic (disperse) dyestuffs for man-made fibres and considerably influences the colour.
3. Groups that induce flatness (planarity) in the molecule as a whole. Substituents promoting planarity are essential in producing useful affinity in dyestuffs for cellulose, cellulose acetates, or polyesters.
4. Lengthy conjugated chains. If a dyestuff molecule is planar, its affinity for cellulose fibres increases regularly with the length of its chain of alternate single and double bonds. To have useful affinity it should have at least eight conjugative double bonds.
5. Phenolic groups. When situated in dyestuff molecules adjacent to certain other groups, phenolic groups confer the ability of combination with a metal mordant.
6. Quinone groups. In the absence of permanent ionic groups, quinones enable a dyestuff to be used for dyeing cellulose fibres by temporary solubilization.
7. Primary amino groups. Primary amino groups on the bases used for azoic dyestuffs can be diazotized and coupled with naphthols (or other materials) already adsorbed on the fibre.
8. Paraffinic chains. Saturated hydrocarbon (paraffinic) chains are used in a few members of several classes of

Table 44: Types of Reactive Groups in Dyestuffs

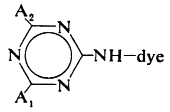
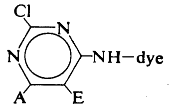
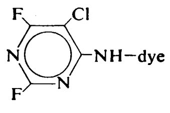
class	chemical structure	
Reactive dyestuffs for cellulose		
Triazine		A_1 = aryl or Cl; A_2 = Cl
Pyrimidine		A = Cl; E = H or Cl
Vinyl derivative	$HO_3SOCH_2CH_2{-}X{-}dye$	X = $-SO_2-$, = $-CONH-$, = $-SO_2NH-$
Reactive dyestuffs for wool		
Acrylamide	$CH_2{=}CHCONH-$(dye-metal complex) $CH_2{=}C(Br)CONH-$dye	
Vinylsulfonyl	$NaO_3SOCH_2CH_2SO_2-$dye	
Trihalopyrimidine		

Table 45: Typical Dyestuffs and Pigments, Natural and Synthetic

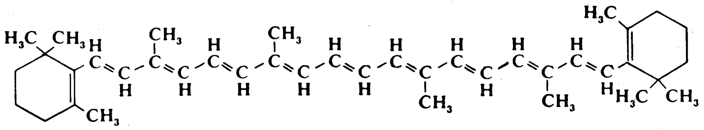
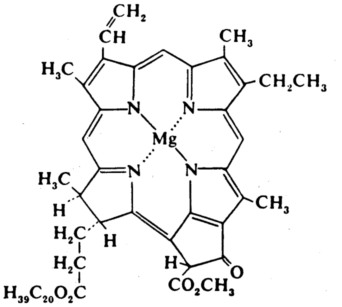
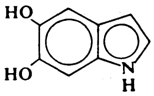
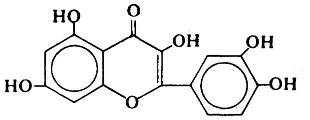
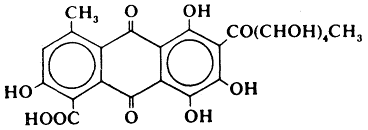
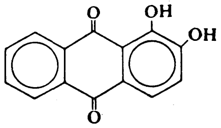
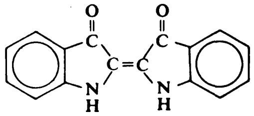
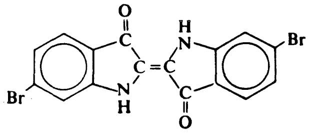
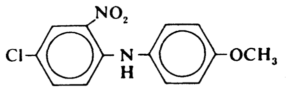
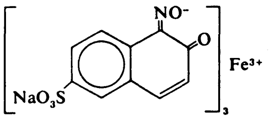
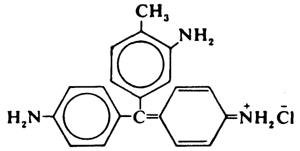
classification and name	*Colour Index* references	chemical structure	uses and properties
Natural biological colorants			
β-carotene	C.I. Natural Yellow 26, C.I. 75130		for fats and oils
Chlorophyll *a*	C.I. Natural Green 3, C.I. 75810		for soaps, oils, perfumes, foodstuffs
Melanin	C.I. Natural Brown 9	 complex; derived from 5, 6-dihydroxyindole	artists' watercolour
Derived biological colorants			
Quercetin (from quercitron bark)	C.I. Natural Yellow 10, C.I. 75670		with mordant, for cellulose fibres or wool
Cochineal (from dried female insects, *Coccus cacti*)	C.I. Natural Red 4, C.I. 75470		with mordants, for wool and silk
Alizarin (from root of madder plant)	C.I. Natural Red 8, C.I. 75330		with mordants, for cellulose fibres, silk, and wool
Indigo	C.I. Natural Blue 1, C.I. 75780, 75790 (C.I. Vat Blue 1, C.I. 73000)		for cellulose fibres and wool
Tyrian purple	C.I. 75800		artists' pigments
Synthetic colorants			
Nitro Amacel Golden Orange III, SRA East Golden Orange III, etc.	C.I. Disperse Orange 15 C.I. 10350		for cellulose acetate and nylon
Nitroso Naphthol Green B, Calcocid Green B, etc.	C.I. Acid Green 1, C.I. 10020	 (as iron complex)	for wool
Triphenylmethane Magenta	C.I. Basic Violet 14, C.I. 42510		for cellulose fibres and miscellaneous materials

Table 45: Typical Dyestuffs and Pigments, Natural and Synthetic (cont.)

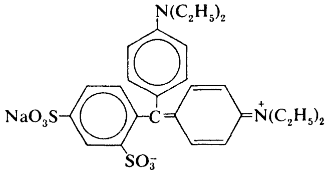
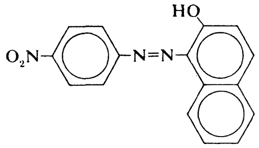
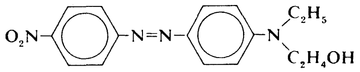
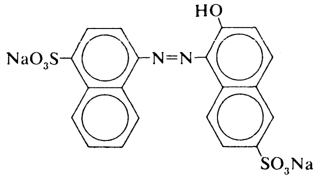
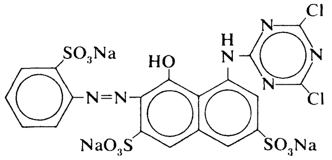
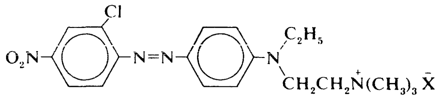
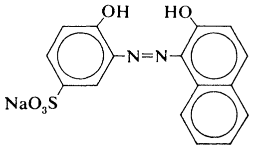
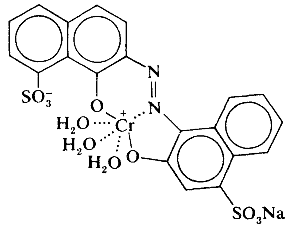
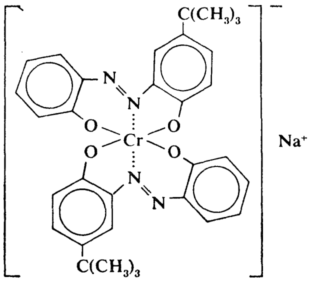
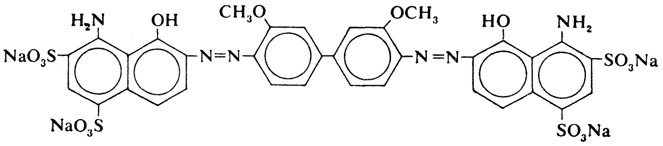
classification and name	*Colour Index* references	chemical structure	uses and properties
Patent Blue V, Kiton Pure Blue V, etc.	C.I. Acid Blue 1, C.I. 42045		for wool, silk, and miscellaneous materials
Azo Para Red, Lake Red, Federal Red, etc.	C.I. Pigment Red 1, C.I. 12070		for miscellaneous materials
Cibacet Scarlet 2B, Celliton Scarlet B, etc.	C.I. Disperse Red 1, C.I. 11110		for cellulose acetates and nylons
Fast Red E, Naphthalene Red EA, etc.	C.I. Acid Red 13 (C.I. Food Red 4), C.I. 16045		for wool and silk (permitted for food in U.K.)
Red reactive dye			for cellulose fibres
Astrazon Red GTL	C.I. Basic Red 18, C.I. 11085		for acrylic fibres
Chrome Violet B, Acid Alizarin Violet, etc.	C.I. Mordant Violet 5, C.I. 15670		with chromium mordant, for wool
Palatine Fast Blue GGN, Ultralan Blue GG, etc.	C.I. Acid Blue 158, C.I. 14880		for wool
Neutral dyeing metal-complex dye			for wool or nylon
Niagara Sky Blue 6B, Direct Sky Blue Green Shade, etc.	C.I. Direct Blue 1, C.I. 24410		for cellulose fibres, silk, and nylon

Table 45: Typical Dyestuffs and Pigments, Natural and Synthetic (cont.)

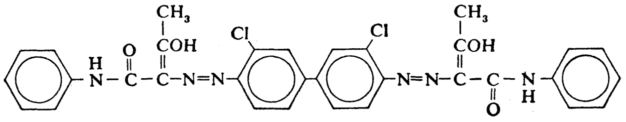
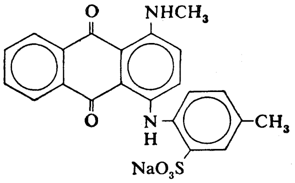
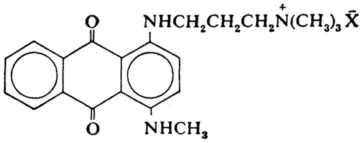
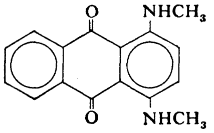
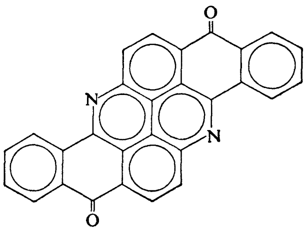
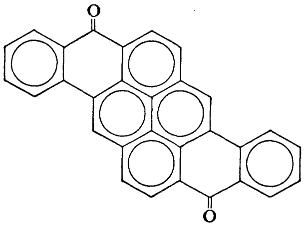
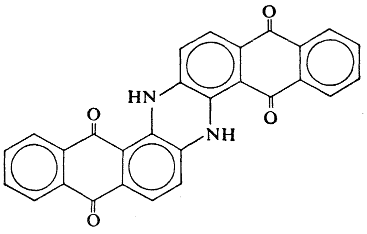
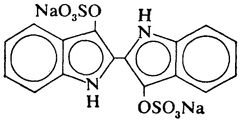
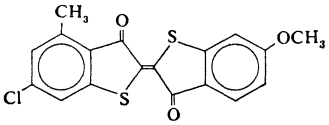
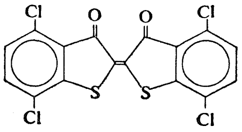
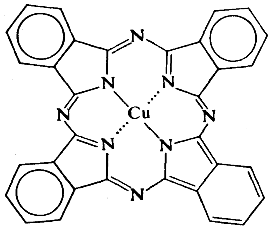
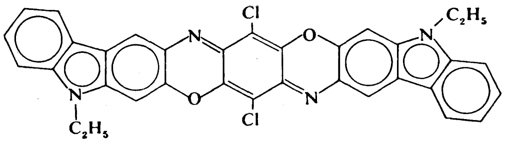
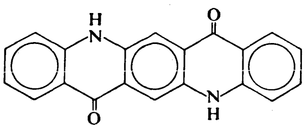
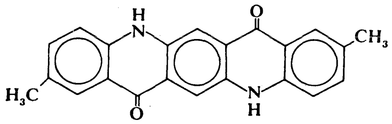
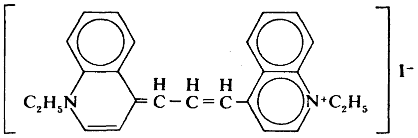
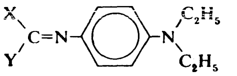
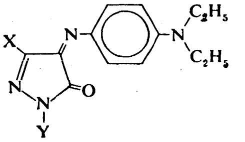
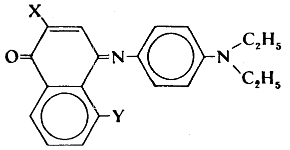
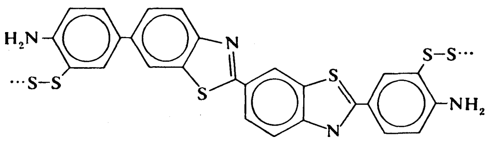
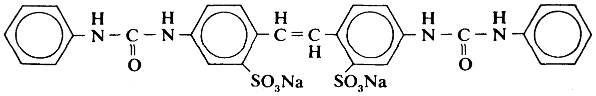
classification and name	*Colour Index* references	chemical structure	uses and properties
Benzidine Yellow	C.I. Pigment Yellow 12, C.I. 21090		good fastness
Anthraquinone			
Kiton Fast Blue 3G, Acilan Astrol B, etc.	C.I. Acid Blue 27, C.I. 61530		for wool and silk
Cationic dye			for acrylic fibres
Cibacet Blue BR, Celliton Fast Blue B, etc.	C.I. Disperse Blue 14, C.I. 61500		for cellulose acetate
Flavanthrone	C.I. Vat Yellow 1, C.I. 70600		for cellulose fibres
Pyranthrone	C.I. Vat Orange 9, C.I. 59700		for cellulose fibres
Indanthrone	C.I. Vat Blue 4, C.I. 69800		for cellulose fibres; also as a pigment
Indigoid and thioindigoid			
Tinosol Blue O, Indigosol O, etc.	C.I. Solubilized Vat Blue 1, C.I. 73002		for cellulose fibres and wool
Ciba Scarlet 3B, Indanthren Scarlet B, etc.	C.I. Vat Red 6, C.I. 73355		for cellulose fibres
Thioindigo Bordeaux			as a pigment; good fastness

Table 45: Typical Dyestuffs and Pigments, Natural and Synthetic (cont.)

classification and name	*Colour Index* references	chemical structure	uses and properties
Phthalocyanine Bermuda Blue, Pigment Fast Blue	C.I. Pigment Blue 15, C.I. 74160		as a pigment; good fastness
Dioxazine Carbazole Dioxazine Violet			as a pigment; good light fastness
Quinacridone Quinacridone Red			as a pigment; good fastness
Quinacridone Magenta			as a pigment; good fastness
Cyanine Kryptocyanine			sensitizer for photographic emulsions
Indoaniline and azomethine			
	. . .		formed in the emulsion during colour development of coloured prints or transparencies (yellow layer)
	. . .		formed in the emulsion during colour development of coloured prints or transparencies (magenta layer)
	. . .	 the X and Y groups have various undisclosed constitutions	formed in the emulsion during colour development of coloured prints or transparencies (cyan-[blue-green] layer)
Sulfur dyestuffs Yellow sulfur dye	. . .	 dyestuffs consisting mainly of this polymeric substance	for cellulose fibres
Fluorescent brightening agents Blancol C Leucophor R, etc.	C.I. Fluorescent Brightening Agent 30, C.I. 40600		for brightening white cellulose fibres (including paper), wool, and nylon

dyestuff. For example, when attached to an acidic molecule for dyeing wool, a paraffin chain of moderate length (containing four to 12 carbons) improves fastness to washing and other wet treatments. Also, when inserted between the polar group and aromatic nucleus of a disperse dyestuff molecule, a short (two- or three-carbon) chain improves the fastness to dry heat. And when inserted between a quaternary ammonium group and an aromatic nucleus, such a chain appears to improve the light fastness of cationic dyes applied to polyacrylonitrile fibres.

In an overall sense, a dye molecule that is either small in total volume, or slim in its cross section, makes rapid dyeing possible because of its ready accessibility to the pores of the fibres.

Formation of chemical bonds

Reactivity. The reactive dyestuffs rely for their retention on the fibre on a principle different from that which controls the dyeing of most other types of dyestuff; that is, they form a covalent (nonionic) chemical bond with the fibre molecules. Such bonds are formed by a reaction between a specific chemical substituent group in the dyestuff molecule with part of the molecule of the fibre.

The reactive groups associated with several classes of dyestuffs are shown in Table 44. Other groups also are employed, and new types of groups are introduced from time to time.

In summary, the molecule of any colouring matter can be viewed as a structure built up step by step, each step designed to give the molecule certain properties and enable it to fulfill certain functions. In Table 45 details of chemical structure that confer the above listed properties on the molecule are presented, and the molecular constitutions of some typical important colouring matters, chosen from nature and from industry, illustrate these principles.

TECHNOLOGY

In this section will be found limited aspects of the classification, manufacture, use, and testing of dyes (for a detailed account of the technology of dyeing, see the article INDUSTRIES, TEXTILE: *Dyeing*).

Classification. Dyestuffs and pigments are classified by chemical constitution, by method of application, by hue, or even by the name of the manufacturer. Generally, the commercial name of a particular colorant consists of: (1) a brand name denoting application class (each one specific to each manufacturer); (2) the name of the hue; (3) suffix letters (often German in origin)—*e.g.,* B, G, or R (blau, gelb, rot); (4) a strength indication—*e.g.,* 150 percent, 250, denoting brands with less diluent (usually salt) than the normal.

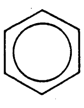
Colour Index references

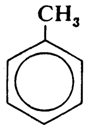
In addition, there is a *Colour Index* which lists (and describes) all commercial dyestuffs by application class and by chemical constitution. For example, the dye Orange II has a *Colour Index* Part I reference as "C.I. Acid Orange 7" and a Part II reference as "C.I. 15510." In the first it is described as a dye for wool, silk, or nylon, used for garment dyeing and for colouring coir and sisal fibres. Also included are details of the grading for fastness to thirteen different agents such as washing, milling, chlorination, light, and so on. Further, the dye is shown as being marketed by about 60 different manufacturers under about 45 commercial names. The Part II reference gives the chemical constitution, method of manufacture, and colour reactions of the dye with common reagents, together with a number of literature references and the information that the dyestuff was discovered by one Z. Roussin in 1876. The manufacturing process for this dyestuff is outlined below.

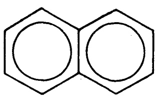
Manufacture. The dyestuff industry uses hundreds of different chemical compounds to produce the thousands of different colorants it markets. These are manufactured by complex and often lengthy series of chemical reactions, utilizing a variety of techniques. The starting materials are relatively few, mainly the aromatic hydrocarbons, of which benzene, toluene, naphthalene, and anthracene are the most important.

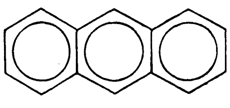
Manufacture of intermediates

From these compounds many hundreds of other substances, known as intermediates, are manufactured. The intermediates usually are not dyestuffs, but from them, ultimately, the dyes are derived. In general, intermediates are formed by one or more of a variety of chemical operations, such as nitration, sulfonation, chlorination, acylation, and so on. Among the most important of the products thus formed are aniline, alkylaniline, acylaniline, chloroaniline, nitroaniline, alkoxyaniline; a large variety of sulfonated aminonaphthalenes and hydroxy-naphthalenes; and aminoanthraquinone and hydroxy-anthraquinone. Further details of the production of dyes are covered in the article INDUSTRIES, CHEMICAL PROCESS: *Dyes.*


benzene


toluene


naphthalene


anthracene

Uses. Although the principal use of dyestuffs is in the dyeing of fibres and fabrics, dyestuffs and pigments are used in large quantities for many other purposes, including the printing of textiles and the colouring of many materials other than textiles, such as leather, paper, and plastics. Pigments are widely employed in paints, varnishes, printing inks, artists' colours, and writing and copying inks. Several colorants are used in foodstuffs, oils, and cosmetics. Dyestuffs also are important as histological stains and for measuring the surface area and porosity of industrial powders.

Special types of dyestuff (chiefly cyanine dyes) are used in making photographic emulsions and others in colour photography. Furs and human hair are dyed by processes involving the application of dyestuffs or related chemical compounds. For foodstuff colouring, specially purified dyestuffs, free from poisonous metals, are used. Only a few dyestuffs are permitted in foods in most countries; these have been rigorously screened, particularly for freedom from carcinogenic properties.

Plastics may be mass-coloured; *i.e.,* the plastic may be ground with a pigment before the final shaping operation, a process that gives an opaque product, or they may be surface-dyed after shaping, by immersion in a dye solution. The latter method does not produce opacity and is suitable for colouring rolls of transparent plastic film or small, transparent articles, such as toothbrush handles.

Standardization, testing, analysis, and identification. After manufacture, water-soluble dyestuffs can be separated from the solution in which they have been prepared by salting out—that is, made insoluble by the addition of salts—after which the dyestuffs are filtered off and dried. Insoluble dyestuffs, which must be easily suspended in water, are finely ground in water with a dispersing agent before drying.

Methods of standardization

Commercial dyestuff products must meet standards of shade, strength, wettability, and so on, so that successive deliveries of any particular material are the same. Even with the best control of manufacturing processes, however, the shade varies slightly from one manufactured batch to another, and the products in their final processing may have to be adjusted to the shade and strength of a standard by addition of little colouring matter of a different hue (but with similar properties) or of a diluent such as salt.

For normal dyeing operations on a technical or a laboratory scale, commercial dyestuffs are used without pretreatment, but for laboratory research purposes it is often necessary to use specially purified dyestuffs, and to check their purity by chemical analysis. Most water-soluble dyestuffs are not as readily purified as are most organic compounds because dyestuffs often are strongly ionic and therefore insoluble in organic solvents. They cannot, therefore, be recrystallized from such solvents. Recrystallization from water often is not satisfactory either because the dyestuff is not sufficiently soluble, or because it does not separate readily on cooling (or if it does, it may do so in a form difficult to remove by filtration). Moreover, most commercial dyes contain large quantities of diluent material, usually common salt, which is difficult to remove completely. For chemical research, therefore, the so-called batch dyestuffs, *i.e.,* dyes obtained directly from the manufacturing plant and unmixed with diluent, are preferable. To purify a water-soluble dyestuff for which a process has not been previously worked out, recrystallization usually is attempted, first from water alone and then, if this is not suitable, from a simple water-miscible solvent, such as acetone, ethanol, propanol, or butanol, or from mixtures of these with water. If all such methods fail, repeated salting out of the dyestuff from water, by additions of sodium acetate, is used, followed by filtration and washing of the precipitate with ethanol, which removes any residual sodium acetate. This procedure gives a product of high purity but in a rather low yield, because much of the dyestuff is lost in the filtrates and in the washings.

Dyestuffs and pigments can be analyzed by the normal methods of analysis used in organic chemistry; *i.e.,* by determination of the total content of the various elements present—carbon, hydrogen, oxygen, sulfur, and so on. With improvement in speed and accuracy, these methods, which now require very small amounts of the test sample, have tended to displace older procedures. They do not, however, detect traces of the coloured impurities, which are often present even after careful purification and which may or may not interfere with the use of the product. Detection of these is best made by the separation method known as chromatography, which is based upon adsorption of the material to be separated on a solid support.

Methods of analysis

Often it is necessary to identify dyes on samples of textile or other coloured material. This usually involves a systematic analysis in which the material is treated in succession with a variety of chemical reagents and solvents, the changes in colour produced or the amount of colour dissolved by the solvent being observed. Such schemes of analysis often are designed merely to identify the dyeing class to which the colouring matter belongs, and this may be all the information needed. Other schemes, however, enable a precise identification of many individual dyes.

(C.H.Gi.)

Alkaloids

Alkaloids, a specific class of chemical compounds found in plants and, occasionally, animals, are notable chiefly for their physiological activity; many have long histories as poisons, narcotics, hallucinogens, and medicinal agents. Generally, alkaloids are basic, or alkaline, substances—*i.e.,* they neutralize acids; their molecules contain chiefly atoms of carbon, hydrogen, and nitrogen, which is the source of their basicity.

Historical background. During the long period when the early art of healing involved magical or religious beliefs, the use of minerals, plants, and animal organs as medicine was not necessarily related to any active constituents they might contain. Not until alchemy evolved into chemistry and medicine became more scientific was it apparent that the medicinal virtues of plant and animal tissues had to depend upon specific compounds in the mixture of components contained therein. Chemists were challenged to separate the constituents and determine by experiment which activities could be attributed to them.

First isolation of an alkaloid

In this regard no drug known to the ancients offered modern research more of a challenge than opium. The first recorded isolation of a crystalline constituent from opium, now thought to have been the alkaloid called narcotine, took place in 1803. In the following year another chemist described what was probably morphine, and a few years later a German chemist purified morphine and recognized its nature as a base. Soon a number of similar substances were recognized and grouped into a new class of compounds known as vegetable alkalies; these later became known as alkaloids.

Progress in the study of the pharmacological properties of alkaloids was stimulated by their potent physiological effects that suggested worthwhile medicinal uses. The chemical study of them, however, was slower and depended upon the growth of organic chemistry, for not until simple organic bases (and acid–base reactions generally) were well understood could the intricate structures of the alkaloids be unravelled.

The first major step came with the study of coniine, an alkaloid from hemlock (a poison known to the ancient Greeks). It had been isolated as early as 1827 and recognized as a derivative of pyridine, an organic base obtained from coal tar. The synthesis of coniine was achieved in 1886—the first synthesis of an alkaloid.

General classification and nomenclature. It is difficult to categorize alkaloids in precise terms. In general it is agreed that an alkaloid must be a product of plant or animal metabolism and must contain at least one nitrogen atom that can be active as a base in acid–base reactions. There are exceptions to this, however—the substance called rutaecarpine, which contains three nitrogen atoms, none of which is basic, is nevertheless classified as an alkaloid because its reactions and the method by which it is synthesized in the plant (biosynthesis) are typical of alkaloids. At the same time, the nitrogenous bases known as purines, of which caffeine is a good example, though they used to be, are no longer treated in works on alkaloids. On the other hand, the substance triacanthine, which belongs to the purine family, is frequently classified as an alkaloid. In principle, alkaloids could be classified either by the biological systems in which they occur or by their chemical structures. In practice both methods are used. For example, it is common to speak of the opium alkaloids because they occur in the opium poppy (*Papaver somniferum*); but it is also common practice to speak of those alkaloids that contain a ring system called indole as the indole alkaloids. This dual classification system actually produces little confusion because there is a rough correlation between the chemical types of the alkaloids and their biological distribution (see below *Biological distribution*).

Basis of alkaloid names

Because most alkaloids were isolated from natural sources before their chemical structures had been determined, they generally bear coined names, which do not denote their chemical structures. Such names may reflect the names of the discoverer (*e.g.,* pelletierine, after the 19th-century French chemist Pierre-Joseph Pelletier), the botanical source (strychnine, isolated from *Strychnos* species), or the biological effect (emetine, an emetic). As a rule, alkaloid names end in -ine, a reference to their chemical classification as amines (see above). In certain cases, alkaloid names are derived in whole or in part from systematic nomenclature systems used in organic chemistry (see above *Organic chemical compounds*). These same systems are encountered in the names of compounds derived from the chemical breakdown, or degradation, products of the alkaloids and synthetic substances related to them. Most alkaloids have one or more of their nitrogen atoms as part of a ring of atoms, frequently called a cyclic system.

In this section alkaloids are grouped into several categories for discussion and, necessarily, many of them appear in each category. These categories are: physiological activity, general chemistry, distribution, classification by structure, and general biology.

GENERAL PHYSIOLOGICAL ACTIVITY

The physiological activity of alkaloids is of importance not only in medicine but also in agriculture and in forensic chemistry. Narcotics addiction and the use of alkaloids as hallucinogens are major social problems. In medicine, alkaloids are employed as narcotics, analgesics, antimalarials, and local anesthetics; as cardiac, uterine, and respiratory stimulants; and as materials that raise the blood pressure, cause the pupils to dilate, or bring about relaxation of the skeletal muscles.

Grouping by physiological effects. A comprehensive survey cannot be made here and, in the following discussion, only the most important of the alkaloids are mentioned.

Poisons. Many alkaloids are of medical importance because they are the cause of occasional poisoning of livestock or man. For example, the senecio alkaloids, and their close relatives isolated from *Crotalaria* and *Heliotropium* species, are only mildly toxic but, in repeated small doses, cause cirrhosis of the liver. The disease is common in livestock in areas where these plants grow, and it has been variously named in the different areas. Another example is the group of ergot alkaloids, produced by a fungus (ergot) that grows on cereal grains; it finds legitimate use in medicine, but the ingestion of cereals

Ergot alkaloids

containing ergot was the source of much serious illness (ergotism) until the cause was well understood. Not everything is known about the chelerythrine-type alkaloids and there has been a suggestion that the ingestion of even very small amounts can induce glaucoma. The alkaloids of the horsetails (*Equisetum*) have come under suspicion in instances of livestock poisoning.

Analgesics. The narcotic of major importance, both in legitimate medicine and in illicit drug traffic, is morphine. Although there are now certain synthetics that substitute for morphine in the relief of pain, this alkaloid is still the most suitable drug for that purpose. Were it not for its addictive properties and the fact that it is sometimes not tolerated well, morphine would be the analgesic and narcotic of choice. Codeine, the methyl ether derivative of morphine, is an excellent analgesic essentially devoid of addictive effects. Diacetylmorphine, best known as heroin, is more powerful as an analgesic but is a strongly addictive drug, which induces in the addict an intense euphoria not obtained with morphine.

Cardiac stimulants. The use of heart stimulants to restore the actions of a malfunctioning heart to normal does not necessarily remedy the causes that underlie the malfunction. The cinchona alkaloid quinidine is the cardiac stimulant used to correct arrhythmias of the auricle (one of the chambers of the heart), and cinchona and rauwolfia alkaloids are used for arrhythmias of the ventricle (the other chamber of the heart). There are no generally acceptable alkaloids for congestive heart failure, in which the pumping action of the heart is inadequate though its rhythm is normal. Generally, another type of material, the digitalis glycosides, is the drug of choice for this kind of heart disease.

Nicotine

Respiratory stimulants. Respiratory stimulants and depressants act on the body by several different mechanisms; owing to the complex nature of these effects, stimulation is often followed by depression, with the result that large doses of stimulants can lead to death by respiratory arrest. As typical respiratory stimulants, nicotine and cytisine show this effect, but they have been used clinically, nonetheless. The alkaloid lobeline is safer in this respect, because eventual arrest of respiration is not as likely, but some alkaloids closely related to lobeline are even more toxic than nicotine. Coniine, too, stimulates respiration, but depression and asphyxia soon follow. Coniine is the main alkaloid from poison hemlock (*Conium maculatum*) and was surely the major ingredient in the hemlock used to kill Socrates. It has been surmised that that poison also contained opium.

Many alkaloids influence respiration; virtually all of them, however, produce other, and often unwanted, effects. Atropine, for example, stimulates respiration in moderate doses, even when respiration has been depressed by morphine, but it also has a number of effects on the brain and dilates the pupils of the eyes.

Certain alkaloids find some use in medicine because they are capable of raising blood pressure. This action can take place in a variety of ways, such as by constriction of the blood vessels or stimulation of the heart or by both at once. Other effects, such as dilation of the pupils, inhibition of gastrointestinal peristalsis, and excitation of the intestines and the uterus, are common.

Ephedrine

Blood vessel constrictors. Ergonovine, one of the ergot alkaloids, finds extensive use in obstetrics to reduce uterine hemorrhage following childbirth, its action being primarily constriction of the blood vessels. Ephedrine also causes constriction of blood vessels and, because of this action, it is widely used to alleviate the discomfort of common colds, sinusitis, hay fever, and bronchial asthma. Certain synthetic agents have now partially replaced the drug for this use. The pupil-dilating action of ephedrine, unlike that of atropine, does not abolish the light reflexes and the accommodation reflexes of the eye. There are a number of other useful dilators and, of these, scopolamine is one of the more powerful. Cocaine, a potent local anesthetic, is also a dilator.

Local anesthetics. Many alkaloids possess local anesthetic characteristics, and some of them do not produce the undesirable effects of cocaine. Extensive experimental modifications of the molecular structure of cocaine have yielded a few compounds that are more effective and less toxic than the parent alkaloid as well as many that are less effective. Modern synthetic local anesthetics, cheaper and often superior to cocaine, have largely, but not entirely, displaced the alkaloid. Although they are not used extensively, the following alkaloids possess acceptable local anesthetic characteristics: derivatives of certain cinchona alkaloids, the erythrophleum alkaloids, an alkaloid from *Fritillaria sewerzowii,* and several derivatives of cotarnine.

Cinchona alkaloids

It is remarkable that only the cinchona alkaloids, especially quinine, have gained and retained prime importance in the treatment of malaria. Extensive modifications of the molecular structures of these alkaloids also have yielded no superior products. An apparent exception to the first generalization is febrifugine, an alkaloid first isolated from *Dichroa febrifuga,* which is a potent antimalarial, especially against bird and monkey infections. Its curative dose, however, is near its toxic dose, and it is not suitable for use in humans. A number of entirely synthetic compounds, however, allow a degree of malaria control that could not be attained with any of the alkaloids.

Muscle relaxants. Relaxation of muscles similar to that produced by the South American arrow poison, curare, is important in medicine, particularly surgery. Useful synthetics are available for this purpose, but the alkaloids of South American species of *Strychnos* and *Menispermum* are much used. These alkaloids, which are the active ingredients of the native curare, are quaternary salts of bisbenzylisoquinolines, the most active being the substances called toxiferine and tubocurarine. They are effective as muscle relaxants only when given subcutaneously or intravenously. On the other hand, dihydro-β-erythroidine, one of the most active tertiary bases from *Erythrina* species, acts as a muscle relaxant only when taken orally.

An important consideration in the therapeutic use of alkaloids is that the physiological actions are generally rather complicated; it is rare that an alkaloid exerts only one type of physiological response. As a result of the study of the relationships between the structural and physiological aspects of alkaloids, it has been possible to direct research into the syntheses of model compounds, which show amplification of one type of activity at the expense of another. It is thus possible in some cases to accentuate desirable effects and minimize unwanted or toxic ones.

Mescaline and LSD

Psychedelics. Many of the so-called psychedelic drugs are alkaloids or their derivatives. Those derived from cacti of the various *Anhalonium* species have long been known and are used in the rituals of some American Indian tribes. Mescaline, the chief of these, causes mainly visual hallucinations. More potent is psilocybin, an alkaloid found in *Psilocybe* species of fungi. Still more potent is LSD, the diethylamide of natural lysergic acid—an alkaloid from ergot—which produces a variety of effects. The dimethylamide of lysergic acid, which is found in some morning-glory (*Convolvulus*) species, is less active.

Morphine and derivatives

Fate of alkaloids in the body. When a toxic substance is presented to a body it may be dealt with in one of three ways: (1) it may be eliminated unchanged in the exhaled breath, the urine, or the feces; (2) it may be chemically altered to a less toxic substance; or (3) it may be conjugated—that is, combined with another substance—to give a product more readily eliminated or tolerated. As an example, morphine is excreted in all three ways: unchanged, chemically altered, and in conjugated form. The conjugate—of unknown nature, though possibly a combination with a sugar molecule—can be hydrolyzed (broken down) to morphine by heating with acid, and it can thus be assayed as morphine following such treatment. For reasons not yet understood, addicts excrete a larger proportion of the conjugate than do nonaddicts. When morphine labelled with carbon-14 in the methyl group attached to the nitrogen atom is given, a portion of the radioactivity appears in the carbon dioxide exhaled from the lungs.

The fate of heroin has not been studied as extensively as morphine, but it is known that there are enzymes in the body capable of removing either or both of the acetyl groups. Since heroin is a more potent narcotic than morphine, it has been assumed that its enzymatic conversion to morphine does not account for its activity. It is possible, nevertheless, that the acetylated substance (heroin) can achieve easier access to the site of its action, where it is then hydrolyzed to the active material.

Small amounts of ingested or injected cocaine are recoverable from urine, but the fate of most of it is not known. When *N,N*-dimethyltryptamine, a naturally occurring psychotomimetic (psychosis-mimicking) drug, is ingested, it is metabolized rapidly and excreted chiefly as indoleacetic acid. At the same time the urine is enriched with 5-hydroxyindoleacetic acid, which may be derived from oxidation of indoleacetic acid or from increased metabolism of serotonin, a substance in the brain that plays a role in the transmission of nerve impulses and may be involved in the action of psychotomimetic drugs.

BIOLOGICAL DISTRIBUTION

The distribution of alkaloids in the plant kingdom is largely confined to the flowering plants, though certain nonflowering forms also produce alkaloids; *Lycopodium,* or club moss, is particularly rich in alkaloids, and a fungus, *Rhizoctonia leguminicola,* produces the alkaloid slaframine. The scent glands of the North American beaver (*Castor canadensis*) contain the alkaloid castoramine, which is structurally related to the nuphar alkaloids found in the yellow pond lilies (*Nuphar* species). The skin glands of a salamander, *Salamandra maculosa,* contain a group of alkaloids that are related to steroids. It is likely that these alkaloids are generated by the animal from steroids in its diet.

Plant families rich in alkaloids

Plants of a number of families are particularly rich in alkaloids. All plants of the poppy family (Papaveraceae) are thought to contain them. In the buttercup family (Ranunculaceae) alkaloids occur in a few genera. Other families containing alkaloids are barberry (Berberidaceae), amaryllis (Amaryllidaceae), pea (Fabaceae), dogbane (Apocynaceae), moonseed (Menispermaceae), custard-apple (Annonaceae), and nightshade (Solanaceae).

The types of alkaloids produced are often characteristic of a particular family or genus. All alkaloids so far obtained from the poppy and custard-apple families are isoquinoline alkaloids (see *Classes of Alkaloids* below for structures and descriptions of this and other alkaloid classes). Although the alkaloids found in plants of the buttercup family are of several types, those from the monkshoods (*Aconitum*) and larkspurs (*Delphinium*) are all of only one type, the aconite alkaloids, and those from the meadow rues (*Thalictrum*) are mostly of another type, the bisbenzylisoquinolines. The cinchonas (*Cinchona*) in the madder family (Rubiaceae) yield mainly the quinine type of alkaloid, whereas the poison nuts (*Strychnos*) in the related Logania family (Loganiaceae) give rise to the highly complex strychnine type. Almost without exception, the alkaloids from the dogbane family are indole derivatives.

Role of alkaloids in plants

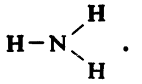
The role that alkaloids play in the life of plants is still largely a matter for speculation. Belladona (*Atropa*) and tobacco (*Nicotiana*) plants produce large quantities of alkaloids, whereas related tomato (*Lycopersicon*) plants do not. When grafts of tomato are grown on stocks of belladona or tobacco, they are almost devoid of alkaloids, whereas the reverse grafts generate fair quantities of them. In the light of these results, the alkaloids would seem to play no essential role in the metabolism of the plants that produce them. That they play some role is suggested by the fact that in many plants prior to seed formation there is a high concentration of alkaloids, which is almost entirely depleted when the seed is ripe. An example of this behaviour is found in the opium poppy (*Papaver somniferum*). In some plants, such as the larkspurs (*Delphinium*), both the seeds and the growing plant are rich in alkaloids. In others (for example, the sweet shrubs [*Calycanthus*]) the seeds are particularly rich, whereas the growing plants contain only traces of alkaloids.

It is thought that the alkaloid content of some plants may protect them from depredation by insects, but there are few, if any plants—even those rich in alkaloids—that are avoided by all insects. *Solanum demissum* seems to be avoided by the larvae of the potato beetle, and several relatively unsuccessful attempts have been made to introduce the genes that control this immunity into other *Solanum* species, especially the potato. Several species of alkaloid-containing groundsel (*Senecio vulgaris* and *S. jacobaea*) are the preferred and possibly the only food for the larvae of the cinnabar moth (*Callimorpha jacobaea*). These larvae have an even higher concentration of alkaloids than the plants on which they have fed; they are not eaten by predators that prey on larvae of related species, presumably because of the alkaloid content.

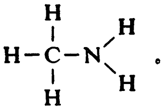
GENERAL CHEMISTRY

Structures of nitrogen compounds

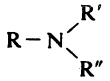
Molecular structure. The structures of the various alkaloid molecules can be understood best by comparison with simpler compounds of nitrogen. In its compounds the nitrogen atom generally forms chemical bonds with three other atoms. In chemical terms, this is to say that nitrogen has a valence of three. For example, the valence bonds of nitrogen in the simple substance ammonia are completed by hydrogen atoms (symbol H, valence one). In representations of molecular structure these bonds are symbolized by lines. Thus, ammonia is denoted by



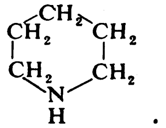
Carbon atoms (symbol C, valence four) also can be joined to the nitrogen atom, giving molecules like that of methylamine, the formula of which is CH_3NH_2:



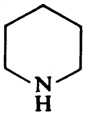
Other combinations of carbon and hydrogen atoms in hydrocarbon chains can be joined to the nitrogen atom, giving the large family of compounds known as amines and represented by the general formula



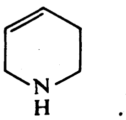
in which R, R′, and R″ are various hydrocarbon groups that may be considered as replacing the hydrogen atoms of ammonia, NH_3; replacement of one, two, or all three of the hydrogen atoms gives primary, secondary, or tertiary amines, respectively. In some amines the hydrocarbon chains are so joined as to form complete rings of atoms. A simple example of this is the substance piperidine, which has the structure



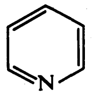
Structures of this type usually are represented diagrammatically by regular polygons in which the corners are understood to signify carbon atoms bearing as many hydrogen atoms as necessary to complete their valences. In this convention, piperidine is represented as



In some hydrocarbons, or hydrocarbon derivatives such as the amines, hydrogen atoms are missing, and the chains or rings are said to be unsaturated or to contain double bonds. Double bonds are indicated in structural diagrams by double lines, as in the formula



Some unsaturated, cyclic compounds contain series of alternating double and single bonds, which give the molecules special properties. An example is pyridine



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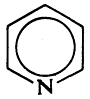
Because these compounds, which are called aromatic, do not behave like typical unsaturated compounds, their structures are frequently drawn with circles within the polygons rather than double lines. In this convention pyridine is written



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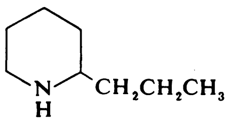
The molecules of the various alkaloids are of a great many different types, which are variations and elaborations of the simple structures discussed above. The alkaloid coniine, for example, is closely related to piperidine (and, therefore, also to pyridine). Its structure is



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Rings in alkaloid molecules

As indicated, most alkaloids contain at least one basic nitrogen atom—that is, a nitrogen atom in an amine-type structure. Generally, also, alkaloids have this or another nitrogen atom in some type of ring system, either aromatic or nonaromatic (aliphatic). But this is not an absolute requirement, for in the molecules of some alkaloids the nitrogen atom is joined to a simple aliphatic hydrocarbon chain. Besides the obligatory carbon, hydrogen, and nitrogen atoms, many alkaloid molecules also contain oxygen atoms, and a small number contain sulfur or chlorine atoms. (See also the section *Organic nitrogen compounds* above.)

Isolation, purification, and analysis. Like other bases, most alkaloids yield salts that are soluble in water when treated with acids, and the separation of these alkaloids from neutral and insoluble plant constituents by salt formation is, therefore, relatively straightforward.

Isolation. Special methods have been developed for isolating commercially useful alkaloids such as morphine. In the main, however, plant and animal tissue is processed in one of three ways to obtain aqueous solutions of the alkaloids. (1) The source material, either fresh or dried, is treated (by a method called extraction that involves the dissolving with reagents of only some components of a mixture) directly with a dilute aqueous solution of an acid that reacts with the alkaloid to produce a soluble alkaloid salt. (2) The material is extracted with an alcoholic solvent, with or without the addition of an acid, after which the solvent is removed by evaporation and the residue—consisting of fats, wax, resins, and other material, as well as the alkaloids—is stirred in contact with a dilute aqueous acid solution that forms soluble alkaloid salts. (3) The material is rendered basic by the addition of aqueous ammonia or sodium carbonate; the mixture is extracted with a water-immiscible solvent, such as ether, and finally, this solution is extracted with dilute aqueous acid. The alkaloids are recovered from the aqueous acid solution prepared by one of the above methods by making the solution alkaline and then extracting it with an immiscible solvent. The last process does not yield alkaloids in which the nitrogen is fixed in a saltlike form; such alkaloids must be precipitated from aqueous solution as complex salts.

Since very few plants elaborate only one alkaloid there still remains the problem of separating and purifying the different alkaloids from the mixture of alkaloids obtained by extraction. The number of constituents, and some indication of their relative abundances, can be learned from analyzing the crude mixture using a system of analysis called chromatography. This is a method of separation that takes advantage of the different degrees of adsorption of the substances to be separated on a solid material, such as alumina or silica. When it comes to actual separation of the alkaloids, chromatography again may be used, or attempts may be made to obtain the alkaloids from solution in crystalline form. The list of solvents from which alkaloids can be crystallized is an extensive one. (The following solvents, in order of increasingly favourable solvent action on alkaloids, are commonly used: hexane, ether, methanol, acetone and ethyl acetate, carbon tetrachloride, dichloromethane, chloroform, dioxane, and pyridine.) There are, of course, exceptions, in which a particular alkaloid is unusually soluble or insoluble in a given solvent, and these can be used to advantage to separate pure alkaloids from the mixtures.

Methods of separation

Separation. Important methods of separation depend upon formation of salts, which often can be crystallized easily. In this case, also, a wide range of solvents is available, including water, methanol, ethanol, acetone, and various mixtures of these solvents with one another and with others, such as chloroform. Picrate salts (salts formed with picric acid) often are used with low-molecular-weight alkaloids, such as nicotine. Picrates crystallize readily and often have sharp melting points, a property that aids in their identification. Hydrochlorides and hydrobromides (salts formed with hydrochloric and hydrobromic acids, respectively) often are easily crystallized, sparingly soluble, and not difficult to purify. Nitrates and oxalates (formed from nitric and oxalic acids), in a few cases, are crystallized easily, often from alcoholic solvents, but they are frequently hydrated (contain water) and seldom have sharp melting points. Those alkaloids that carry a phenolic hydroxyl group (an $-OH$ attached to an aromatic ring) have acidic properties; they form water-soluble salts in the presence of strong bases and therefore are separated easily from their relatives lacking phenolic groups.

Purification. Distillation and sublimation (vaporization of a solid) in a vacuum are practical methods for separating alkaloids of moderate molecular weight. Countercurrent distribution, a separation technique that depends on repeated partition between an immiscible solvent and an aqueous phase at controlled acidity, is a useful procedure for separating certain alkaloids, but it is laborious unless an automatic instrument for carrying out the process is available.

Determining the amount of individual alkaloids present in a sample is important in the drug trade and in forensic practice. Reliable procedures for the analysis of frequently encountered alkaloids have been developed and are much used.

Combustion analysis

Structure determination. The molecular formula of an alkaloid, which gives only the kinds and numbers of atoms present in the molecules, can be determined by combustion analysis—that is, by burning the alkaloid to end products, such as carbon dioxide and water, which can be collected and weighed. The functional groups (active groupings of atoms) present must be detected by specific chemical tests, characteristic of each group.

Mass spectrometry. A modern technique known as high-resolution mass spectrometry breaks down the molecule into fragments and accurately measures the mass of each. This not only provides a precise molecular formula but at the same time reveals details of how the atoms are joined in the molecules. Many types of alkaloids have been examined by this technique, and the fragmentation patterns exhibited by the various types are now well understood.

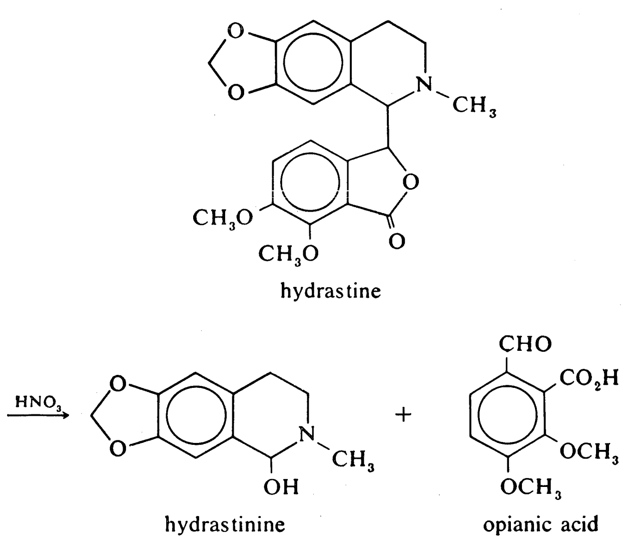
Spectroscopy. Infrared spectra (patterns of absorption of light in the infrared region) give information about the presence or absence of a number of functional groups, chief of which are the carbonyl, hydroxyl, and amino groups. Ultraviolet spectra (light absorption in the ultraviolet region), give valuable information about aromatic systems and unsaturated groups conjugated with aromatic rings or carbonyl groups.

Nuclear magnetic resonance. A measure of the nuclear behaviour of the atoms and detailed information on the nature and environment of the hydrogen atoms may be derived by nuclear magnetic resonance; such information frequently tells much about the structure of the molecule as a whole.

X-ray diffraction. The X-ray is an extremely powerful tool for structure determination. It depends for its success on the interpretation of patterns obtained when X-rays are diffracted by the atoms in crystals. Preferably carried out with a crystalline salt or other derivative containing a heavy atom, such as bromine or iodine, it also has been used with alkaloids containing as elements only carbon, hydrogen, oxygen, and nitrogen. There are several advantages to the method: if suitable crystals can be grown, a complete study can be carried out with very small amounts of material; the method requires no previous knowledge of the functional groups present; and, most important, it can be used in cases in which chemical degradation does not work (see below).

Oxidation

Chemical methods. The structure of an alkaloid can be determined by chemical reactions that break the intact molecule into fragments, the structures of which are already known or can be determined. In ideal cases, all of the carbon and nitrogen atoms can be accounted for in the various fragments, which result either from a single type of chemical assault on the molecule or from the use of a number of breakdown reaction paths. Oxidation of the alkaloid is the most direct attack and, in selected cases, gives enough information to deduce the total structure of the original alkaloid. Upon oxidation with partly diluted nitric acid, the alkaloid hydrastine, for example generates the known compounds hydrastinine and opianic acid, as shown in the equation:

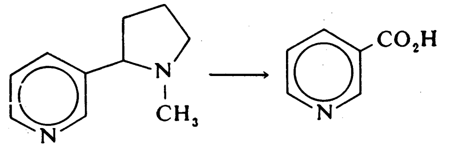


hydrastine

hydrastinine

opianic acid

Together these compounds account for all the carbon, nitrogen, and oxygen atoms of the original alkaloid and permit its hypothetical reconstruction. On the other hand, oxidation often destroys much of the molecule. Nicotine, for example, upon oxidation with a variety of oxidizing agents gives mainly nicotinic acid, the remainder of the molecule being converted into small fragments of no value in deducing the structure:



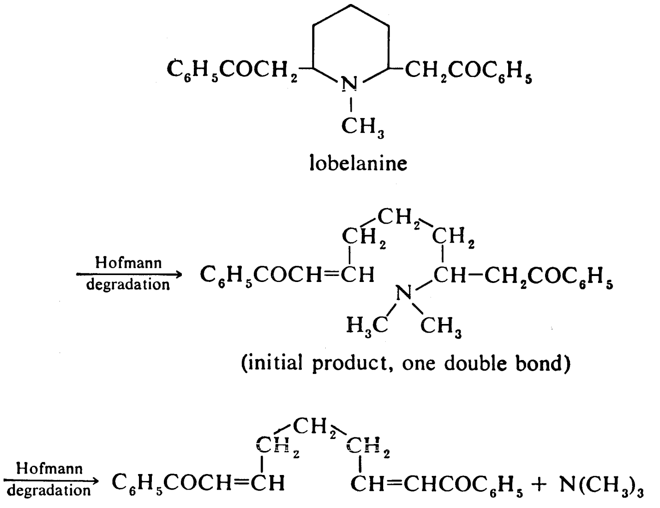
nicotine

nicotinic acid

Strychnine (see below) upon oxidation, again wih a variety of oxidizing agents, gives a number of oxidation products, all of which provide useful structural information, but only upon further degradation.

Hofmann degradation

To obtain structural information in greater detail, it is convenient to use the method known as the Hofmann degradation. This procedure, developed by a German chemist, August Wilhelm von Hofmann, consists of treating the alkaloid with methyl iodide to convert the basic nitrogen atom into its quaternary, or salt, form; the resulting quaternary iodide is then subjected to strong alkali. Under favourable conditions, any ring containing a nitrogen atom will be opened by this procedure, with the introduction of a double bond into the molecule at the point where the nitrogen atom was attached. At the same time, the nitrogen atom remains part of the molecule as a methylamine derivative. A repetition of the process may generate a second double bond and a new tertiary base, but this occurs only when the nitrogen atom in the original alkaloid is part of two rings. Otherwise the second Hofmann degradation gives a methylamine and a nitrogen-free compound containing two double bonds. These reactions are illustrated by the following equations:



lobelanine

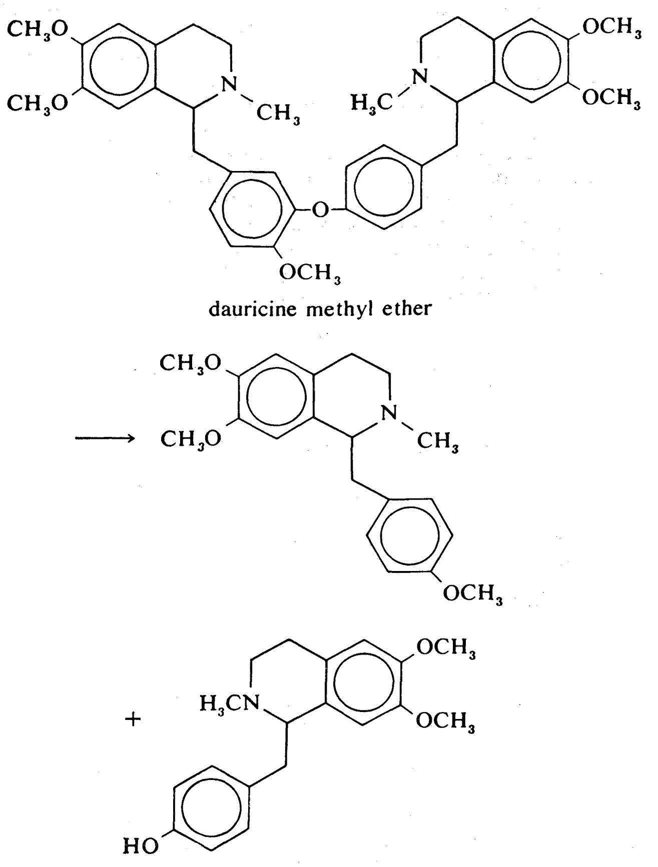
(initial product, one double bond)

(final product, two double bonds) trimethylamine

If the amine formed in the second-stage Hofmann degradation is trimethylamine, then the original alkaloid must have contained an *N*-methyl group. If the second Hofmann degradation has not eliminated nitrogen from the molecule, a third-stage degradation is sometimes possible, leading to a triply unsaturated carbon compound. There are many alkaloids that do not lend themselves to Hofmann degradation, however, and the reaction may fail at any of the stages. Nevertheless, when it can be employed, this method of degradation is especially useful, because it generates unsaturation in the molecule, which may then open the way to oxidative degradation. A related process achieves the same result as the Hofmann degradation, but with concomitant reduction of the double bond. It is carried out by reacting the methiodide with sodium amalgam or occasionally by reduction with sodium borohydride or with hydrogen in the presence of a catalyst.

Frequently, Hofmann degradation and subsequent oxidations proceed satisfactorily, but the final product, an acid, is difficult to identify because of its complexity.

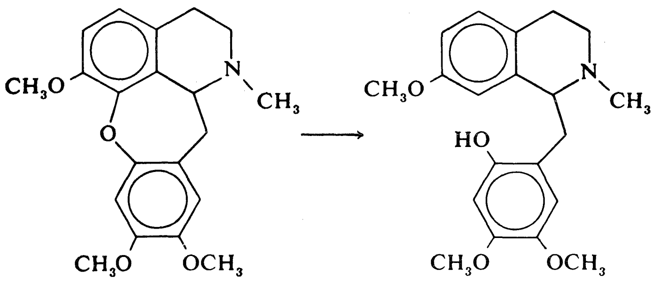
Special cases. *Bisbenzylisoquinolines.* Such is particularly the case with the bisbenzylisoquinoline alkaloids. In addition to relatively simple and identifiable phthalic acids, these alkaloids give rise to extremely complex acids (such as the alkoxyphenylether dibasic acids), which are difficult to identify because their syntheses are so difficult. A reaction is available that eliminates such problems by dividing the molecule into two approximately equal parts at the diphenyl ether junction. For example, when the methyl ether of the bisbenzylisoquinoline alkaloid dauricine is dissolved in ether or tetrahydrofuran and added to a freshly prepared solution of metallic sodium in ammonia, it gives rise to two simple benzylisoquinolines that are easily identifiable. The cleavage reaction is depicted as follows:



dauricine methyl ether

(easily identifiable benzylisoquinoline products)

The somewhat simpler alkaloid cularine was the first to be subjected to this reaction, and the product, also a benzylisoquinoline, was relatively easy to identify:

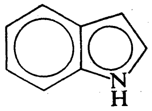


cularine

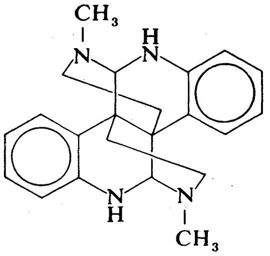
(benzylisoquinoline product)

The results of degradation reactions must always be interpreted with caution.

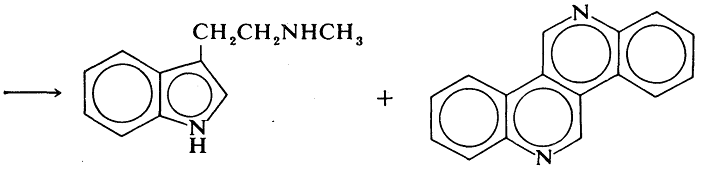
Calycanthine. An example of misleading information obtained by degradation is found with the alkaloid calycanthine. Attempts to degrade the alkaloid by the Hofmann procedure were unsuccessful but, by chance, it was found that benzoylation of the alkaloid yielded a product that was easily identified as a benzoyl derivative of *N*-methyltrypamine. This strengthened the suspicion that calycanthine was an indole derivative, like *N*-methyltryptamine, as already suggested by a colour reaction, which is given usually only by indole compounds. In a further experiment, however, calycanthine was heated to a high temperature wih zinc dust in an inert atmosphere, giving *N*-methyltryptamine, as well as a product that was later named calycanine:



indole



calycanthine



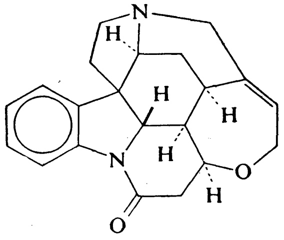
N-methyltryptamine

calycanine

Surprisingly enough, the structure of calycanine was shown by synthesis not to be an indole derivative. An X-ray diffraction study of the calycanthine molecule showed that its structure did not include indole nuclei, and it was concluded that the extreme conditions of the pyrolysis, which could have induced great structural changes, had only eliminated molecular fragments, whereas the mild benzoylation reaction had actually brought about deep-seated molecular rearrangements.

Strychnine. A similar, though still more complicated, example of misleading results of degradative experiments is presented by investigations of the strychnine molecule. Elucidation of this extraordinary complex structure of overlapping rings drawn as

Determining the structure of strychnine

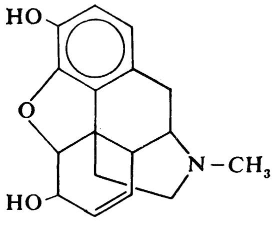


strychnine

(bonds shown - - - extend below the plane of the paper, – above it)

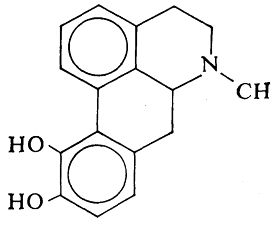
is often cited as a major triumph of the classical techniques of degradative chemistry. For more than a century the structure remained a puzzle. That the structure yielded finally to attack was due chiefly to the skill of a British chemist, Sir Robert Robinson, and an American chemist, Robert Burns Woodward. Strychnine had been subjected at an early date to pyrolysis under a variety of conditions, from which a number of degradation products had been identified. When the structure was finally known, it was recognized that proper juxtaposition of these fragments was sufficient to write the complete structure of the alkaloid.

Morphine. The determination of the structure of morphine, which is



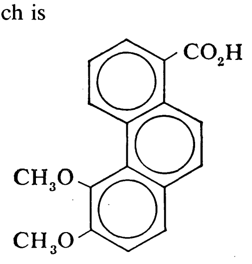
morphine

was thwarted by rearrangements that occurred when the molecule was subjected to chemical attacks. A partly reduced phenanthrene skeleton (composed of three fused six-membered rings) was postulated for the alkaloid because phenanthrene was obtained when the alkaloid was heated with zinc dust. Other degradations led to hydroxyphenanthrenes, and heating the alkaloid to 130° C (266° F) with strong hydrochloric acid generated the substance called apomorphine, the structure of which



apomorphine

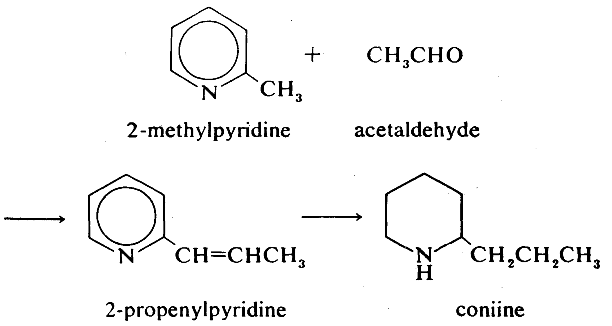
was ascertained by reactions that have since become standard. Furthermore, on Hofmann degradation, the dimethyl ether of the alkaloid gave a vinylphenanthrene, which on oxidation was converted to the corresponding acid, the structure of which is



(acid product)

The structure of this compound was proved by synthesis. Nonetheless, structures of morphine, based upon that of these products—apomorphine, in particular—were untenable because the nature of the rearrangement that gave apomorphine was not understood. A great deal of degradative and synthetic chemistry was necessary before these misleading results could be reinterpreted and the finally accepted structure could be written.

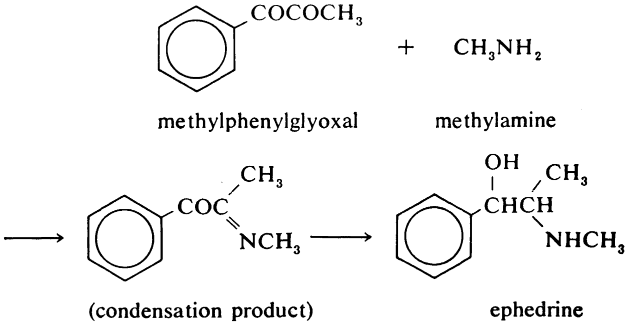
Synthesis. As mentioned above, the first synthesis of an alkaloid was that of coniine. In this synthesis, acetaldehyde was condensed with 2-methylpyridine to give 2-propenylpyridin; on reduction with sodium in ethanol this substance gave synthetic coniine:



The synthetic product is not quite identical with the natural alkaloid because it does not exhibit the optical activity of that substance. Optical activity is a property of materials that is only detected when plane polarized light is passed through a crystal or solution of the substance; the observed effect is the rotation of the plane of polarization of the light (as determined in an instrument called a polarimeter). Optical activity results when molecules are capable of existing in two isomeric forms that are related to one another like mirror images, or right- and left-hand gloves. These forms are designated *dextro (d)* or *levo (l),* according to whether they rotate plane polarized light to the right or left, respectively. The molecules of many alkaloids, like those of other biological materials, exist in nature in only one of the optically active forms. Synthetic processes, however, usually produce mixtures of both isomers, which are called *dl–* forms. Separation of the forms, a process called resolution, often can be achieved by crystallizing the *dl–* mixture with another substance already optically active. In the case of coniine, the natural material is the *d–* form, and resolution of the synthetic mixture is carried out by crystallizing the salt formed with *d*–tartaric acid.

Resolutions of *dl–* mixtures

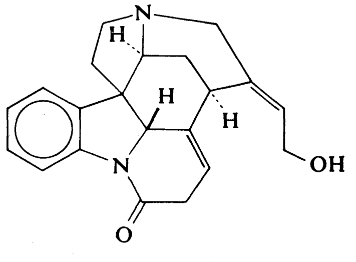
The alkaloid ephedrine, used as a nasal decongestant, has been synthesized by several methods; the most direct of these, a process used commercially, consists in condensation of methylphenylglyoxal with methylamine and reduction without isolation of the initial condensation product. The sequence of reactions is as follows:



In this case, too, it is necessary to resolve the product into *d* and *l* forms, only one of which (in this case, the *l* form) corresponds to the natural substance. The *l–* and *d–* mandelic acids, respectively, proved to be satisfactory for this purpose, each combining with the oppositely rotating form of the mixture to produce a salt that can be purified by crystallization. Conversely, the *d–* and *l*–ephedrines can be used to prepare pure *l–* and *d*–mandelic acids, respectively. The examples of coniine and ephedrine illustrate the general principle that the synthesis of an alkaloid usually necessitates a resolution of the forms, either as a final or intermediate step.

Synthesis of strychnine

The synthesis of the poisonous alkaloid strychnine posed as great a challenge to chemists as did the determination of its structure and, for the same reason, the extreme complexity of the molecule. The final synthesis of the alkaloid was reported in 1954. Complicating the problem are six asymmetric carbon atoms in the strychnine molecule (the spatial arrangement of which is shown above). No less than 15 sequential reactions were necessary to achieve the synthesis of isostrychnine of the correct configuration:



isostrychnine

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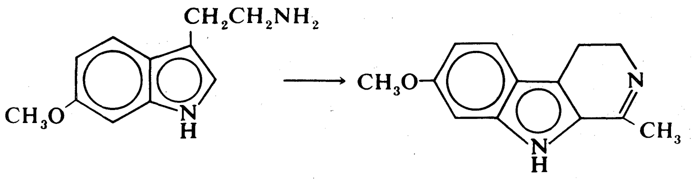
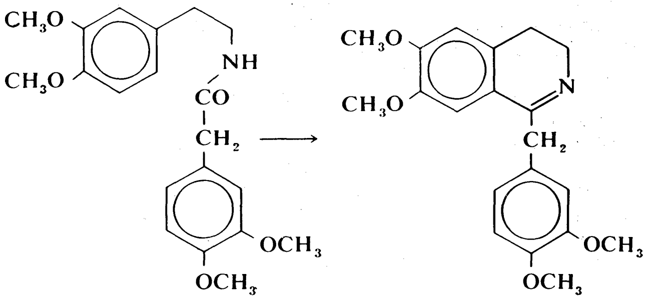
It was necessary to insure that the correct stereochemistry was achieved at each step; in one instance optical resolution of an intermediate was needed. Since isostrychnine can be prepared from strychnine, and since it can be reconverted to the alkaloid, synthesis of it represents a synthesis of strychnine itself.

Of chiefly historical interest is the relatively simple alkaloid harmaline; it was the first alkaloid of the indole type the structure of which was determined, and the first indole alkaloid to be synthesized. Since the initial synthesis, there have been a number of others. Typically these begin with the substance 6-methoxytryptamine:


6-methoxytryptamine harmaline

Quite analogous syntheses of 1-substituted isoquinolines can be carried out beginning with β-arylphenethylamines. A typical example is the preparation of dihydropapaverine from N-(3,4-dimethoxyphenethyl)-3,4-dimethoxyphenylacetamide:

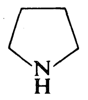

N-(3,4-dimethoxyphenethyl)-3,4-dimethoxyphenylacetamide dihydropapaverine

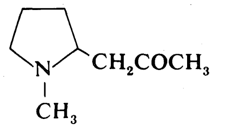
Conversion of the dihydro- derivative to the alkaloid papaverine by dehydrogenation can be accomplished by a variety of procedures.

CLASSES OF ALKALOIDS

It is convenient to classify alkaloids according to their structural types, taking into consideration the nature and number of their nulei (rings of atoms) and the substituents attached to these. The simplest, aside from purely aliphatic bases that contain no rings, are the alkaloids the molecules of which include the five-membered pyrrole ring composed of one nitrogen and four carbon atoms. The more complex structures vary considerably in the types of ring systems present and in the number of nitrogen atoms they contain.

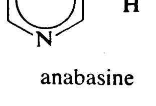
Pyrrolidines. The unsubstituted compound, pyrrolidine itself, has been found in the carrot, *Daucus carota,* and in tobacco; its N-methyl derivative has also been isolated from tobacco and from belladonna. Hygrine, found in the leaves of the cocoa plant, is among the few other pyrrolidine alkaloids.

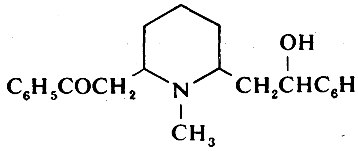

pyrrolidine

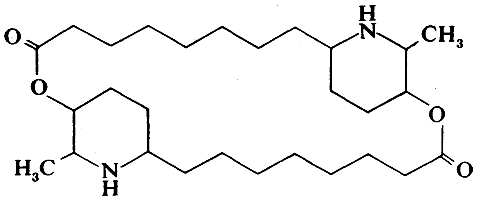

hygrine

Pyridines. The pyridine alkaloids constitute a large group in which the pyridine nucleus is present in fully aromatic, partly reduced, or fully reduced form. Those alkaloids in which the pyridine ring is fused to a second ring are discussed under other classes. The best known of the pyridine alkaloids is nicotine, which has both a pyridine and a pyrrolidine nucleus. Anabasine is formed by the union of two pyridine nuclei. The important lobelia and sedum alkaloids, of which there are well over 20, are mostly 2-monosubstituted and 2,6-disubstituted piperidines, with or without a methyl group on the nitrogen atom. The substitutes are of the type $-CH_2COR$ or $-CH_2CH(OH)R$, in which R may be methyl ($-CH_3$), ethyl ($-C_2H_3$), or phenyl ($-C_6H_5$). Lobeline is the chief alkaloid of the family used in medicine. The alkaloid carpaine from papaya (*Carica papaya*) had long been given the formula $C_{14}H_{25}O_2N$ and assigned a structure that included a 13-membered lactone ring. Mass spectral data, however, showed that the correct formula is $C_{28}H_{50}O_4N_2$ and that the alkaloid contains a double lactone ring of 26 atoms, as shown in the structural formula:

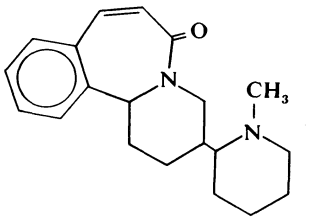
Lobeline


anabasine

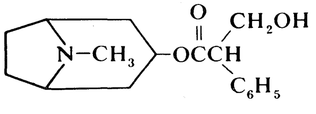

lobeline


carpaine

There are many possible ways of combining several pyridine nuclei, as well as forming extra rings and adding substituents (see below). Astrocasine from *Astrocasia phyllanthoides* is one example.

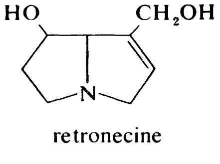

astrocasine

Tropanes. Bridging of the 2,6- position of piperidine (fully reduced pyridine) by a two-carbon unit gives the nucleus of the tropanes. Tropinone (see below *Emulation of biosynthesis*) is the keto derivative of this nucleus, and tropine is the corresponding alcohol. Esterification of tropine with tropic acid gives the alkaloid atropine, one optically active form of which is hyoscyamine:

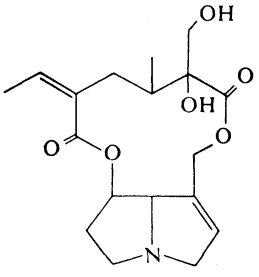

atropine (hyoscyamine)

Pyrrolizidines. The ring system formed by the fusion of two five-membered rings with a common nitrogen atom is known as pyrrolizidine. The base itself has not been found to occur naturally but some important derivatives are widespread. The most common and most important of these is retronecine, which occurs in an esterified form

Retronecine

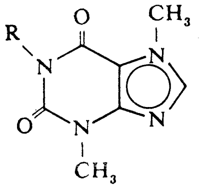

retronecine

in a number of genera in the families Compositae, Leguminosae, and Boraginaceae. These esters are commonly referred to as senecio alkaloids because of their abundance in *Senecio* and closely related genera. More often than not esterification is by a dibasic acid, often containing isoprene units, so that large ring systems may be present. Those alkaloids, such as retrorsine, that are cyclic diesters of acids with branched chains are more toxic than straight-chain diesters. These alkaloids are mutagenic in fruit flies (*Drosophila*) and can cause liver tumours in rats. A number of senecio alkaloids occur in plants as amine oxides, and these are equally toxic. Retrorsine has the structure:



retrorsine

Purines. Although the purine bases (with two rings in the molecules, five- and six-membered, respectively, each including two nitrogen atoms) are nitrogenous compounds elaborated by many different plant species and genera, many of them are not considered alkaloids because of their major biochemical roles. Caffeine, having the following structure,



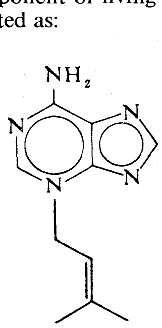
caffeine ($R=CH_3$)
theobromine ($R=H$)

however, is generally classified as an alkaloid and is the most widespread of the purine alkaloids. Aside from its presence in tea and coffee, it occurs in a number of species of holly (*Ilex*) and in many other plants.

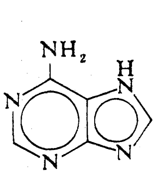
Caffeine

The stimulating effects of caffeine are well known and doses larger than one gram in man may cause alarming symptoms though there seem to be no recorded fatalities from overdoses. Dogs are more sensitive to this alkaloid, and poisoning by theobromine, a closely related alkaloid, has been noted often when cacao meal has been used as food for pigs and chickens. Such meal may contain up to 5 percent of theobromine, which can be recovered and converted commercially into caffeine by methylation with methyl chloride in the presence of caustic soda.

The purine nucleus also is present in the alkaloid triacanthine, which occurs in *Gleditsia triacanthos,* and which is remarkable because of its high nitrogen content. It is cleaved by strong acids to adenine, an extremely important component of living cells. Triacanthine and adenine are depicted as:



triacanthine



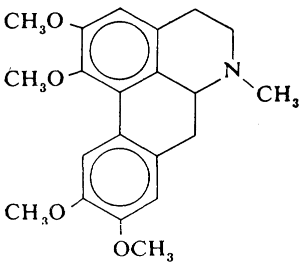
adenine

Isoquinolines. The isoquinoline nucleus consists of two fused six-membered rings, one of which contains a nitrogen atom. In most isoquinoline alkaloids there are substituents on the 1- position of the isoquinoine nucleus and the pyridine ring is in its tetrahydro- form. Most frequently the 1- substituent is a benzyl group, although it may be other alkyl, aryl, or arylethyl groups. Furthermore, the 1-benzylisoquinolines undergo a series of transformations, with or without the addition of carbon fragments, to generate various polycyclic types.

Morphine and codeine

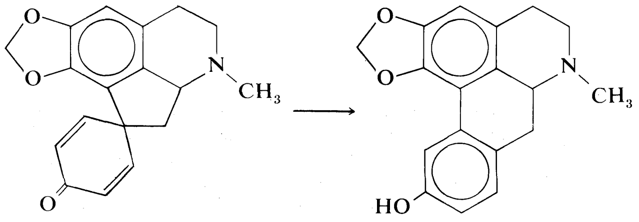
The isoquinoline alkaloid classes represented by papaverine, protoberberine, and protopine have already been mentioned. Morphine (see above) and codeine do not, on superficial examination, belong here, but alkaloids found with them in the plants are benzylisoquinolines; they themselves may be transformed into benzylisoquinolines; and they arise biogenetically from benzylisoquinolines. The structures of morphine and codeine were long obscured by the observation that morphine is converted into apomorphine when heated with strong acid. The alkaloid sinomenine has a similar nuclear structure but lacks the oxygen bridge.

Apomorphine is only one representative of a large variety of alkaloids known as aporphines. These differ from apomorphine only in the position, number, and type of substituents on the nucleus. Glaucine, first isolated from the horn poppy (*Glaucium flavum*) and subsequently found in many other plants, was the first aporphine to be synthesized. Pharmacologically the aporphines are not important. Glaucine has the structure:



glaucine

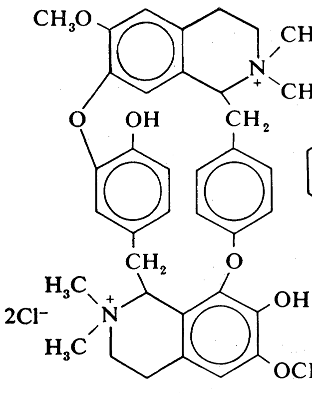
The proaporphine group of alkaloids was more recently discovered. They occur in *Croton* (Euphorbiaceae), *Nelumbo* (Nymphaeaceae), *Stephania* (Menispermaceae), *Ocotea* and *Neolitsea* (Lauraceae), and *Papaver* and *Meconopsis* (Papaveraceae). They are prepared synthetically by phenolic oxidation of benzylisoquinolines, and their biogenetic origin seems to be similar. There is evidence that they are precursors of some aporphines. Mecambrine has the typical nuclear structure of these alkaloids, which differ from one another in the number and position of substituents and the degree of unsaturation of the quinone nucleus. When mecambrine is heated in dilute hydrochloric acid it undergoes rearrangement to mecambroline.



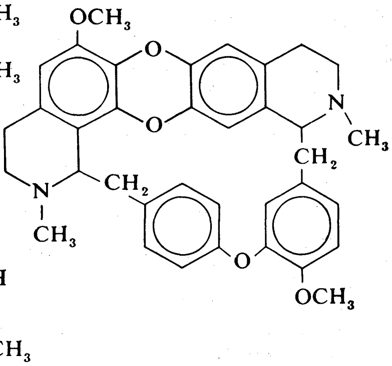
mecambrine mecambroline

Bisbenzylisoquinolines

The bisbenzylisoquinolines also are generated biologically by oxidation of benzylisoquinoline—in this case, two being incorporated per molecule. These alkaloids also occur in a number of genera belonging to several plant families. The simplest representative is dauricine, found in some *Menispermum* species. There is much variation in the structures of this group of alkaloids. Tubocurarine, already mentioned for its use in surgery, is a quaternary base containing two diphenyl ether linkage in a large ring; isotrilobine is an example of the alkaloids of this family that have three diphenyl ether linkages.

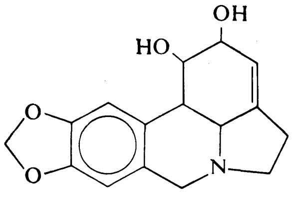


tubocurarine chloride

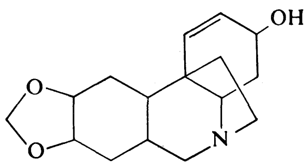


isotrilobine

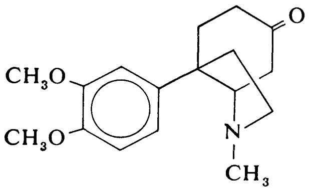
A rather distinct class of isoquinoline alkaloids is represented by the numerous alkaloids of the Amaryllidaceae. Included are a variety of nuclear structures, that of lycorine being representative of one group. Crinine is a variant of this type and an "open" form of this structure is represented by mesembrine, an alkaloid from *Sceletium* species (Ficoidaceae). The last named base still retains the polyhydroindole system present, sometimes in disguised form, in all of these alkaloids, the structural formulas of which are:



lycorine



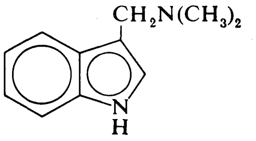
crinine



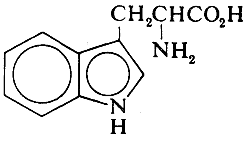
mesembrine

Gramine and strychnine

Indoles. Gramine, a simple indole derivative, and strychnine, a far more complex one, already have been mentioned. The alkaloids of the indole family, without exception, are derived biogenetically from the amino acid tryptophan, and with only one exception, namely that of gramine, they retain the aminoethyl chain of tryptamine.



gramine



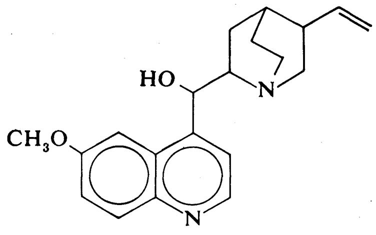
tryptophan

These alkaloids have been isolated from plants representing well over 30 families, the Apocynaceae, Rubiaceae, and Loganiaceae being particularly rich sources. These alkaloids often evince powerful physiological action, and one particularly complex group has shown promise in the control of leukemia. Vincaleucoblastine, one of some 10 alkaloids present in *Catharanthus roseus* (*Vinca rosea*) being used clinically to treat leukemia, is a representive of one of the indoline-indole dimeric alkaloids.

The determination of the structure of such a complex molecule required the combined techniques of organic chemistry and spetroscopy, the study of light absorption and its relation to molecular structure. Even alkaloids containing only one indole nucleus have presented challenges to the structural chemist.

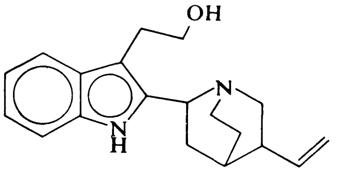
Quinine

Quinolines. The quinolines, too, constitute a large group of alkaloids. Remarkably enough, they are also derived from tryptophan. The biogenesis of quinine, which has the following structures,



quinine

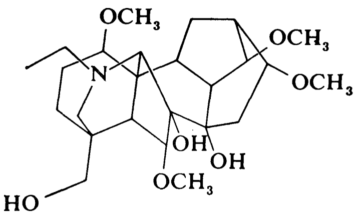
is not understood in detail but it is known that the indole nitrogen of tryptophan remains in the quinoline portion, and that the side chain nitrogen appears in the quinuclidine moiety. There is some evidence that cinchonamine, which is present along with quinine, is an intermediate in the synthesis of quinine.



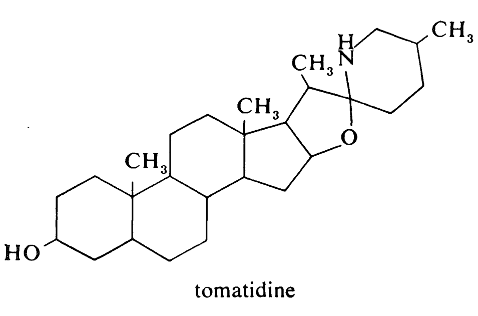
cinchonamine

The alkaloid calycanthine (seen above), which for long had been regarded as an indole alkaloid and can be readily converted into indole derivatives, also has its origin in tryptophan.

Terpenoids and steroids. Most of the terpenoid and steroid alkaloids can be viewed as derivatives of the nonnitrogenous terpenoids and steroids. These are common constituents of many living systems and are derived from mevalonic acid. In these alkaloids a nitrogen atom from ammonia presumably has been incorporated into the terpenoid or steroid molecule. The alkaloids also are distributed widely in plants (and are found occasionally in animals); the family Ranunculaceae yields many terpenoid bases, such as lycoctonine, and the Solanaceae yield largely steroid alkaloids, as represented by tomatidine.



lycoctonine

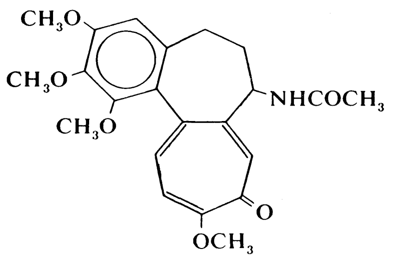


tomatidine

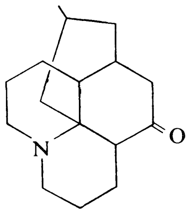
Miscellaneous. Since plants do not elaborate alkaloids for the convenience of classification, many alkaloids do not fit neatly into classes. All of those that have been subjected to scrutiny with labelled precursors, however, fall into one or another of the recognized pathways of biosynthesis.

Colchicine

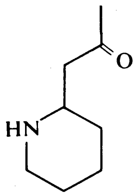
Among the miscellaneous alkaloids, colchicine, from the autumn crocus (*Colchicum autumnale*), is of importance not only because it has been used to cure gout but because it has been the cause of many deaths following accidental ingestion. Colchicine was one of the first compounds known to bring about profound chromosomal changes in plant cells. The lycopodium alkaloids, typified by lycopodine, are represented by over 30 known examples. They appear to be derived from two isopelletierine units. The quinolizidine alkaloids, commonly found in *Lupinus* species and other leguminous genera, are typified by sparteine.



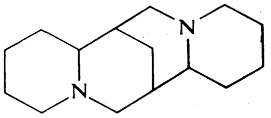
colchicine



lycopodine

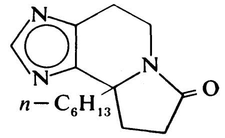


isopelletierine



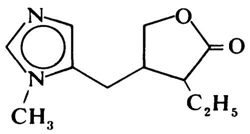
sparteine

Glochidicine is one of three closely related alkaloids found in *Glochidion philippicum* (Euphorbiaceae).

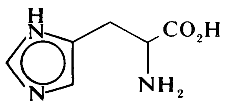


glochidicine

It is of interest because, like pilocarpine, which is used in medicine as an agent to induce perspiration, it is one of the few alkaloids derived from the amino acid histidine:



pilocarpine



histidine

In addition to the alkaloids mentioned here there are many others that are of interest because of their molecular structure and physiological activity.

GENERAL BIOLOGY

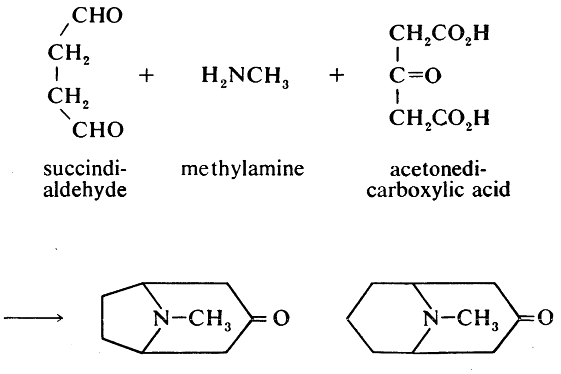
Biosynthesis. It has long been suspected that the nitrogen atoms and some of the carbon atoms of alkaloids have their origin in amino acids, nitrogen-containing organic compounds that are constituents of proteins and other biological materials. A proposal describing in detail how certain alkaloids could have been derived from amino acids was made in 1917. When the radioactive isotope of carbon, carbon-14, was made generally available, it became possible to introduce radioactively labelled amino acids into plants and verify that the label appeared in the alkaloid under investigation, confirming the incorporation of the amino acid into the alkaloid. Even earlier, indirect evidence had been obtained to explain the structures of the various alkaloid types on the basis of their origin from different amino acids. It became feasible to select the more plausible of alternate structures of a new alkaloid on the basis of the possibility of origin of the structure from known amino acids.

Derivation from amino acids

Emulation of biosynthesis. Attempts to emulate the presumed biological synthesis (biogenesis) of an alkaloid by in vitro experiments often met with unqualified success. These experiments were conducted with compounds known or thought to be present in the plant and were carried out under conditions that might prevail in the plant cell, namely, aqueous neutral solutions maintained at room temperature. The first such successful experiment was a spectacular synthesis of the alkaloid tropinone achieved simply by bringing together succindialdehyde, methylamine, and acetone in aqueous solution.

In vitro processes

The yield was improved when acetonedicarboxylic acid was substituted for acetone, decarboxylation of the product taking place either spontaneously or by heating in acid solution. Extension of this synthetic scheme to pseudopelletierine required the replacement of succindialdehyde by its homologue glutardialdehyde.

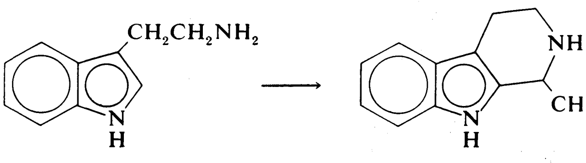


$$\begin{matrix} CHO \\ CH_2 \\ | \\ CH_2 \\ CHO \end{matrix} \quad + \quad H_2NCH_3 \quad + \quad \begin{matrix} CH_2CO_2H \\ | \\ C{=}O \\ | \\ CH_2CO_2H \end{matrix}$$

succindialdehyde methylamine acetonedicarboxylic acid

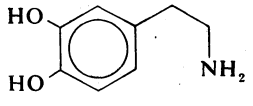
tropinone pseudopelletierine

Subsequently, many successful syntheses of alkaloids of varied structural types were achieved under equally mild conditions. Tryptamine, an indole base derived from the amino acid tryptophan, in the presence of acetaldehyde readily yielded tetrahydroharman:

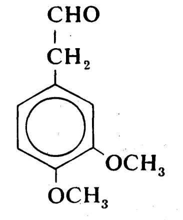


tryptamine tetrahydroharman

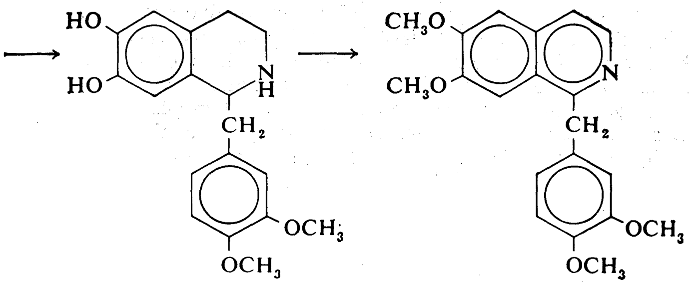
Similarly, 3,4-dihydroxy-β-phenethylamine and 3,4-dimethoxyphenylacetaldehyde yielded a tetrahydroisoquinoline, which on methylation of the oxygen atoms and dehydrogenation was converted to papaverine:



3,4-dihydroxy-β-phenethylamine



3,4-dimethoxyphenylacetaldehyde



(a tetrahydroisoquinoline) papaverine

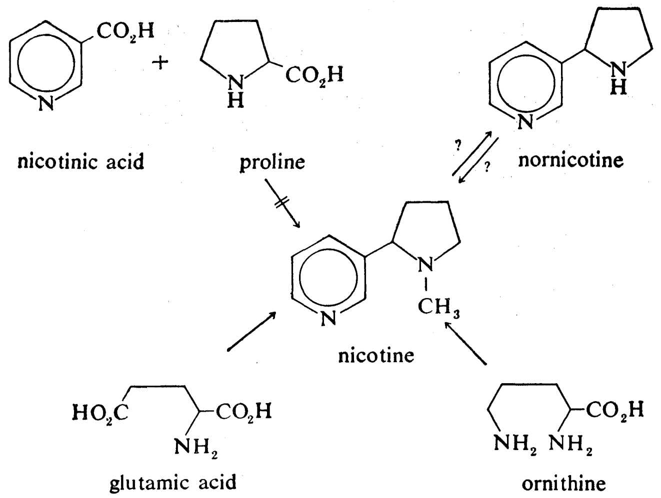
The pyruvic acid, of which the above aldehyde is a decarboxylation product, underwent the same reaction. This acid could be prepared from the amino acid tyrosine. The phenethylamine involved in the reaction is a decarboxylation product of dihydroxyphenylalanine, which also is known to be derived ultimately from tyrosine.

Radioactive labelling and biosynthesis

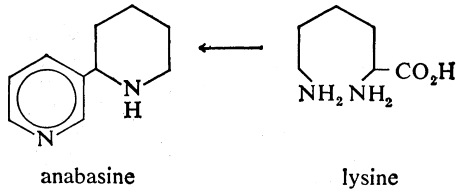
Use of tracers. With the aid of carbon-14-labelled substances, it became possible to test biogenetic hypotheses by direct experiments and, with a few exceptions, results were as expected. In many cases labelling experiments provided additional detail lacking in the original biogenetic hypotheses. In the meantime the processes by which the precursor amino acids are synthesized in the cell from carbon dioxide and ammonia also were being determined.

As early as 1912 it had been surmised that nicotine arose biologically from a combination of nicotinic acid with the amino acid proline. Experiments involving living tissues of *Nicotiana* species, however, have shown that two other amino acids, glutamic acid and ornithine, are incorporated into the five-membered ring of nicotine instead of proline. It proved difficult to trace the exact sequence of reactions leading to nicotine, and it is still unclear whether the nornicotine (nicotine missing its methyl group) is a precursor or a conversion product of nicotine.

The relationships among these compounds are shown in the following scheme:



The biogenesis of a structurally related alkaloid, anabasine, also is subject to some uncertainty, although the amino acid lysine is known to be incorporated into the saturated six-membered ring of the compound:

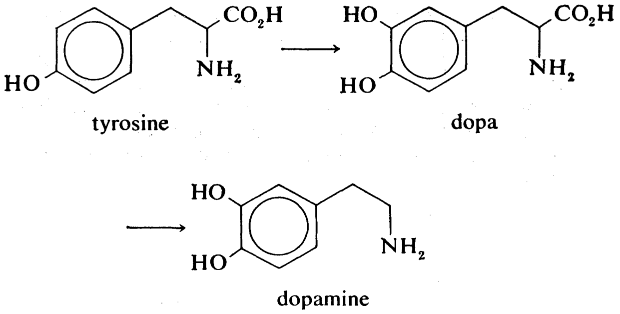


A particular advantage of labelling techniques is that they permit one to determine not only what precursors are involved in the biosynthesis of a given product but also the way in which the precursors are used—that is, the specific atoms of precursor molecules that become atoms in the molecules of the products. This information is secured by synthesizing precursor molecules bearing labelled atoms at known positions in the molecule, introducing them into the organism and, after an appropriate time period, isolating the product from the organism. If, as expected, the product bears the radioactive label, it is carefully degraded chemically, each atom being isolated in turn so that it can be determined whether or not it is radioactive. The correspondence reveals precisely what molecular transformations have taken place. Such experiments are, of course, extremely laborious; however, they yield information that can be obtained in no other manner.

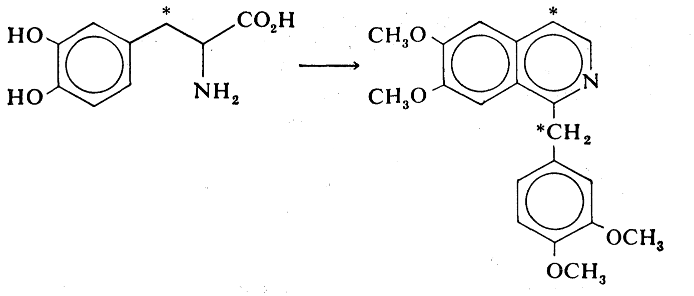
Biosynthesis with duplicate amino acids

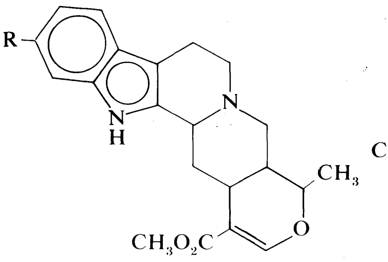
Radioactive labelling experiments also have been useful in unravelling the tangled situations that arise when two molecules of a single amino acid are involved in the biosynthesis of a given alkaloid molecule. In such cases the two units may go through separate pathways of transformation, with the ultimate result that they are incorporated in different forms, and possibly to different extents, in the alkaloid molecule. On the other hand, both units may be incorporated into the alkaloid in an identical form or, after incorporation, they may form a symmetrical intermediate; in either of these latter cases the alkaloid will show an identical degree of incorporation of the two units. The relative degree of incorporation of two amino acid units into an alkaloid structure is best shown by radioactive labelling. An example of this situation and the procedure used to clarify it is given by the amino acid tyrosine, two molecules of which are believed to be incorporated into the molecules of a number of alkaloids (papaverine, morphine, narcotine, thebaine, hydrastine, berberine, and chelidonine, to name a few). Furthermore it is known that tyrosine is transformed in plant cells into a substance called dopa (for dihydroxyphenylalanine), which also is believed to be incorporated into these alkaloids to the extent of two units for every alkaloid molecule. A further transformation product of dopa (called dopamine), however, is incorporated only into one part of the molecules of these alkaloids. The progression from tyrosine to dopamine is represented as follows:

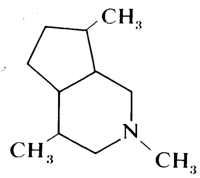
Dopa



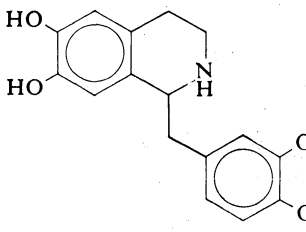
It seems, therefore, that there must be two separate routes by which dopa is incorporated into the alkaloid molecules. This conclusion was verified by the finding that radioactively labelled dopa was incorporated into papaverine in such a way that the two labelled carbons were both radioactive but to different extents.

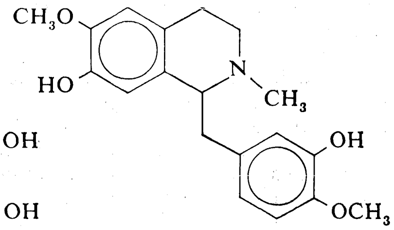

labeled dopa labeled papaverine


ajmalicine (R = H)
reserpinine (R = OCH_3)


skytanthine

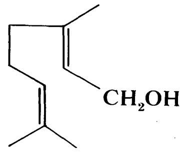
It seems likely that the substance norlaudanosoline, or a close relative of it, is the primary condensation product of two molecules (which may or may not be identical) derived from dopa, and the related base reticuline is actually known to occur in a number of plants that elaborate alkaloids of this type. With the aid of labelling experiments, it has been verified that reticuline is indeed the precursor of the simple benzylisoquinolines and their subsequent biological descendants, which are many of the alkaloids under discussion. Indeed, it has been shown that some of these alkaloids are intermediates in the synthesis of others in living plants.

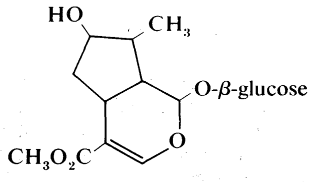

norlaudanosoline

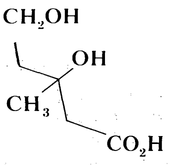

reticuline

It should be emphasized that the incorporation of amino acids into biosynthetic pathways involves more often than not a partly broken down (catabolized) amino acid in which one or more of the functional groups had been either eliminated or altered. Long before biosynthetic studies with labelled compounds became possible the degradation of amino acids into a variety of catabolites by living cells had been recognized.

Special cases. There are, however, a large number of alkaloids whose biosynthesis cannot be explained solely by using the degradation products of amino acids. Many of the indole alkaloids, for example, such as ajmalicine and reserpinine, include as part of their molecular structures a 10-carbon fragment derived from naturally occurring hydrocarbons, the monoterpenes, like geraniol or loganin. Loganin itself occurs in the Madagascar periwinkle *Catharanthus roseus* plant (*Vinca rosea*). Mevalonic acid, the source of the five-carbon-atom (isoprene) unit in such other natural products as the terpenes and steroids, is the precursor of geraniol and loganin. Radioactive labelling experiments have shown that two units of mevalonic acid are incorporated into skytanthine, an alkaloid that has been isolated from the South American plant *Skytanthus acutus*. The structures of these related compounds are shown below.

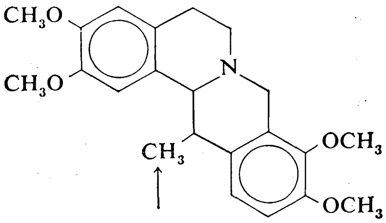

geraniol

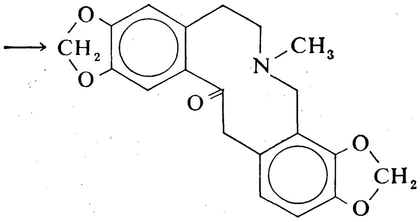

loganin

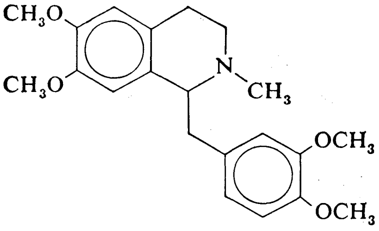

mevalonic acid

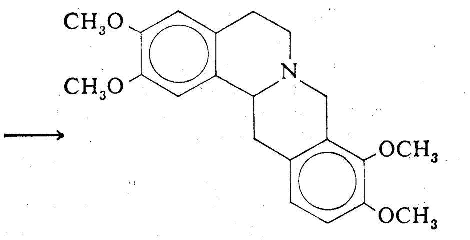
Origin of methyl groups

Many alkaloids contain methyl ($-CH_3$) groups joined to oxygen or nitrogen atoms and, in virtually all instances studied so far, this methyl group has originated (joined to a sulfur atom, as $S-CH_3$) in methionine [$CH_3-S-CH_2-CH_2-CH(NH_2)-CO_2H$], an amino acid that occurs in free form in plant sap and in combined form in many proteins. In addition, methyl groups attached to carbon atoms also come from methionine in certain alkaloids, corydaline, for example. The carbon atoms of the methylenedioxy ($-O-CH_2-O-$) groups of protopine also originate from methionine. It has been shown that the methylenedioxy group is formed by a process in which one of the oxygen atoms is methylated first and then, with removal of two hydrogens, ring closure with the adjacent oxygen atom takes place. The mechanism of this reaction is unknown, and it has not been possible to duplicate it in the test tube. A somewhat similar example of the removal of hydrogen atoms with consequent ring closure, also not duplicated in the laboratory, takes place in the biosynthesis of tetrahydropalmatine from *N*-methyltetrahydropapaverine, represented in the formulas:


corydaline
C-methyl group, coming from methionine, indicated by arrow


protopine
methylene groups coming from methionine indicated by arrows


N-methyltetrahydropapaverine



tetrahydropalmatine

(R.H.M.)

Isoprenoids

Isoprenoids are members of a class of organic compounds formed in plants and animals by combination into larger molecules of units containing five carbon atoms arranged in the characteristic pattern present in the simple substance isoprene, from which the class takes its name. The name terpene, or terpenoid, derived from turpentine, which is a mixture of isoprenoids, is also applied to this class of compounds. In plants, isoprenoids occur in the essential oils, in the gummy exudates (oleoresins and latices) of many trees and shrubs, as substances affecting growth (such as the hormone gibberellic acid), and as red, yellow, and orange pigments (carotenoids). Chlorophyll, the green pigment essential in photosynthesis, is partly isoprenoid, as are certain alkaloids, nitrogen-containing compounds present in many plants. In animals, isoprenoids comprise various oily or waxy substances as fish-liver oils, wool wax, and the yellow pigments in egg yolk, butterfat, feathers, and fish scales.

Several isoprenoids are vitally important in metabolic processes in animals: vitamins A, E, and K are wholly or partly isoprenoid in molecular structure, as are the ubiquinones (coenzymes Q), which are involved in the derivation of energy by the oxidation of food. In insects, certain isoprenoid substances influence maturation and mating behaviour or are used to communicate alarm or to repel predators. Steroids, a class of compounds of great importance in both plants and animals, are not isoprenoids but are derived directly from them. Among many isoprenoids that are valuable in industry and commerce, turpentine, rosin, camphor, menthol, and natural rubber may be mentioned.

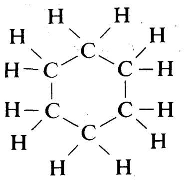
Studies of the isoprenoids

The problems in molecular structure presented by the isoprenoids have challenged the imagination and skill of organic chemists since the latter part of the 19th century. The first studies were mainly concerned with the structures of the monoterpenes, the molecules of which contain 10 carbon atoms, but, as those structural patterns became familiar and techniques for investigation were developed, attention was turned increasingly toward those isoprenoids containing 15 to 40 carbon atoms. Otto Wallach, a German chemist, recognized in 1887 that the numerous ways in which the 10 carbon atoms are arranged in the molecules of the monoterpenes, or the 15 carbon atoms in the sesquiterpenes, all could be regarded as resulting from various combinations of a fundamental unit that contains five carbon atoms that are always connected in one of several possible ways. Wallach's proposal, called the isoprene rule, was of great help in understanding the structures of the more complex members of the class. By the mid-20th century, emphasis had fallen upon the origin of isoprenoids in biological systems, their relationships to other classes of natural products, and their powerful physiological effects in plants and animals.

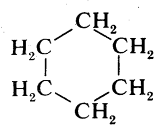
CHEMISTRY OF ISOPRENOIDS

The six-carbon ring

Structural features. A structural feature especially common among isoprenoid compounds is a ring of six carbon atoms; the simplest compound possessing this structure is cyclohexane (not an isoprenoid), which is represented by the structural formula **1**, by a condensed version **2**, or simply by the hexagon **3**. In compounds of this kind, the six ring atoms are not coplanar, but the ring usually is puckered, as shown in **4** and **5**. Molecules that can be regarded as formed by replacing one or more hydrogen atoms of cyclohexane (although few of them can actually be prepared in this way) by other atoms or by radicals (groups of atoms) can exist in different forms, depending on which hydrogen atoms are replaced. In the isoprenoid alcohol menthol, for example, three of the hydrogen atoms of cyclohexane have been replaced, each by a different radical: structure **6** shows only the bonding pattern without implying the spatial arrangement, but **7** and **8** specify the orientation of the bonds and the conformation of the ring in the natural form of menthol.



1



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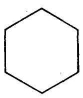

3

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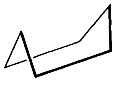

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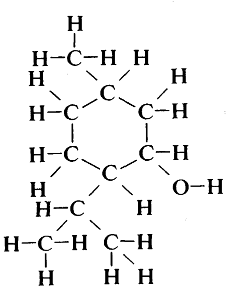
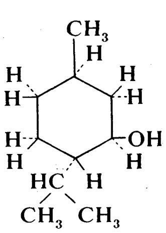


6



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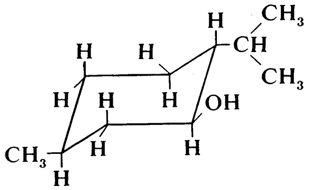
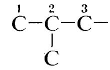


8

Carbon skelton

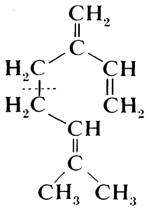
The term carbon skeleton is used to describe the pattern in which the carbon atoms are bonded together in a molecule, disregarding atoms of other elements and differences between single and multiple bonds. Most chemical reactions of organic compounds do not break bonds between carbon atoms and therefore leave the carbon skeleton unchanged. In many isoprenoids, rings of three, four, or five carbon atoms form part of the molecular structure. In these small rings, the bond angles are severely distorted from the values typical of stable molecules, and the carbon–carbon bonds can be broken with unusual ease. Many reactions in which carbon skeletons are rearranged were first observed during investigations of the isoprenoids and caused considerable confusion until the causes of their occurrence were recognized.

Classification. The isoprenoids are broadly classified according to the number of isoprene (C_5H_8) units they contain, and they range in size from volatile oils of molecular formula $C_{10}H_{16}$ to giant molecules such as that of natural rubber, which contains about 4,000 isoprene units. The classes are: monoterpenes, $C_{10}H_{16}$; sesquiterpenes, $C_{15}H_{24}$; diterpenes, $C_{20}H_{32}$; triterpenes, $C_{30}H_{48}$; tetraterpenes, $C_{40}H_{64}$; polyterpenes, $(C_5H_8)_n$. Many of the isoprenoids possess carbon skeletons that may be regarded as built up from isoprene units



$$\overset{1}{C}-\underset{\underset{C}{|}}{\overset{2}{C}}-\overset{3}{C}-\overset{4}{C}$$

linked "head to tail"; that is, carbon atom 1 of one unit is bonded to carbon atom 4 of the next. Formation of additional bonds in a variety of ways leads to monocyclic, bicyclic, and further subclasses in which one, two, or larger numbers of rings are present. *β*-myrcene (**9**), an acyclic monoterpene; limonene (**10**), a monocyclic monoterpene; *α*-pinene (**11**), a bicyclic monoterpene; and vitamin A (**12**), an oxygenated monocyclic diterpene, exemplify this further classification; the dotted lines in the structural formulas indicate the division of the carbon skeletons into isoprene units. The structures of most triterpenes and tetraterpenes show that they were formed by establishment of a tail-to-tail bond between two smaller units: in the structural formula of the important triterpene hydrocarbon squalene (**13**), for example, the arrow indicates the bond uniting two sesquiterpene portions.



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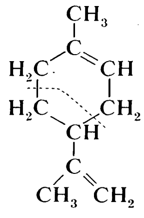

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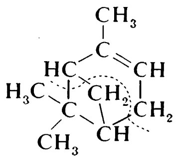

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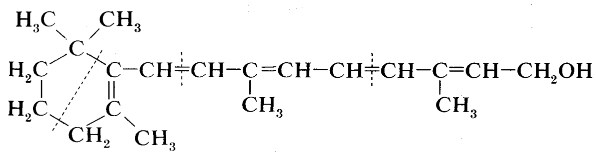

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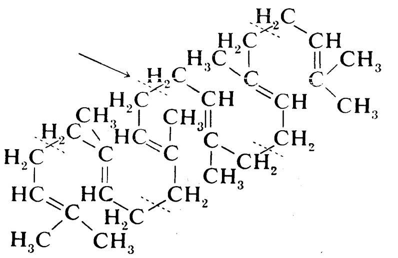


13

Natural sources. Isoprenoids are not uniformly distributed among plants or animals, but certain classes of these compounds are typical of broad groups of tissues. The essential oils—more or less volatile materials obtainable from odorous plants by physical means, such as extraction with solvents, distillation, or expression—are most often monoterpene or sesquiterpene hydrocarbons or their oxygenated derivatives, such as alcohols, aldehydes, or ketones. Menthol, citral, camphor, limonene, and α-pinene are examples of this group; often a single species or group of related species is characterized by the presence of one of these compounds. The nonvolatile substances present in resins produced by trees of the pine family contain diterpene carboxylic acids belonging to three types, abietic, palustric, and elliotinoic; the latices of a few species of plants contain the polyterpene hydrocarbons rubber or gutta-percha.

Essential oils

The livers of fishes and other animals are particularly rich in oils that are largely acyclic triterpenoid hydrocarbons, especially squalene. Tetraterpene hydrocarbons and a few of their oxygenated derivatives make up most of the carotenoid pigments that are widespread in plants and animals.

Importance. *Biological functions.* The natural biological functions of isoprenoids are, in many cases, unknown. Those that have been deduced, however, are extremely diversified. In both plants and animals, isoprenoids are intermediates in the formation of steroids, which, in animals, have been shown to perform numerous essential tasks. In plants, tetraterpene carotenoid pigments are closely associated with chlorophyll (a partially isoprenoid compound) in photosynthesis; the same group of pigments are, in animals, the source of vitamin A, a substance essential to visual processes. Vitamin A is also involved in growth, reproductive function, and neural development in animals. Other vitamins that are wholly or partly isoprenoid include vitamin E, important in reproduction, and vitamin K, necessary for the blood-clotting process.

The role of the monoterpenes and sesquiterpenes produced by plants has not been established, although it has been suggested that they attract certain insects and repel others. This idea is connected with the presence in insects of some of the very same compounds, which are used to signal alarm, to repulse predators, to mark the way between food sources and the nest, to attract members of the opposite sex and stimulate mating. Certain plants produce isoprenoids that are very similar to hormones involved in the development of insects that prey on those plants; the plant substances prevent the maturation of the insect, thus serving to defend the plant. The rosin acids of pine trees and the rubber and gutta-percha in latices of various plants may serve as wound-sealing agents.

Uses. Of the uses that man has found for isoprenoids, many were established in antiquity, as ingredients of perfumes and incenses, flavourings and spices, and varnishes and medicinals. Amber, a fossilized isoprenoid resin, has been prized as a gemstone since prehistoric times. Rubber was in use by the Central and South American Indians before the Spanish conquest.

Modern applications are extremely diversified. Turpentine, long employed as a solvent, is now mostly used as a source of its individual components, which are raw materials for chemical processing. Products derived from turpentine include ingredients for perfumes, vitamin A, lubricant additives, insecticides, resins used in adhesives, and industrial chemicals. Rosin, usually modified by chemical treatment, is widely used to make inexpensive soaps and coating materials.

Isolation and identification. Numerous procedures have been developed for the isolation of isoprenoids from their natural sources. The selection of a technique is influenced by such factors as the chemical and physical properties of the compound and its abundance and distribution in nature. Volatile and plentiful substances as turpentine are obtainable by distillation of oleoresins; rosin acids and fatty acids occur together in tall oil, a by-product obtained in the manufacture of paper pulp from pine wood, and they are separated by fractional distillation at reduced pressure. Extremely rare compounds, as the insect hormones, have been isolated by chromatography (a method of separating the components of a solution, or mixture, based on selective adsorption onto a layer or column of suitable material). For the isolation of heat-sensitive perfume ingredients from flower petals, the laborious process of enfleurage is employed: the petals are placed in thin layers of carefully purified fat, in which the floral oils dissolve; they are recovered from the fat by washing with alcohol.

Isolation techniques

Chemical identification of isoprenoids, like that of organic compounds in general, falls into a classical pattern of steps that has the goal of furnishing a complete description of the molecule. This description includes specification of the number and kinds of the atoms present, the chemical bonds that hold the atoms together, and the three-dimensional arrangement of these atoms and bonds. The sequence of operations necessary for such an identification comprises purification, determination of atomic composition, and assignment of structure. Verification of the assignment is accomplished by synthesis of the compound from others of previously established structure by chemical reactions that proceed in known ways.

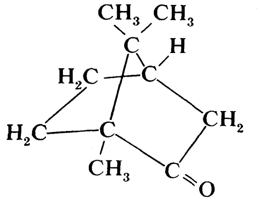
Purification. Isoprenoids can be purified by various techniques that depend primarily upon the physical properties of the compound, as melting or boiling points or solubilities in various other substances. The chemical properties may be an advantage or a disadvantage in the application of these methods: a compound in which heating induces a chemical reaction would be destroyed rather than purified by a procedure requiring the application of heat, as distillation or sublimation. On the other hand, methods depending on physical properties often cannot effect a satisfactory separation of a mixture of compounds possessing very similar physical properties; sometimes such mixtures can be converted by chemical treatment into a mixture of new substances that can be more readily separated; the separated components may then be reconverted into the original compounds.

In general, solid compounds can be purified by recrystallization; and volatile compounds (either solid or liquid), by distillation. Nonvolatile liquids or solids contaminated by very similar substances can be purified by chromatography. Because even small amounts of contaminants can give rise to seriously misleading results in chemical identification, purification of a compound for this purpose must be carried to the highest attainable degree; it is highly desirable to subject a substance to a succession of purification procedures based upon different principles, as distillation and crystallization. The rigour of this requirement is far higher than in purification of a substance for the purpose of suiting it to typical practical applications, for which it is often sufficient only to remove components that would be objectionable in the intended application (*e.g.,* those that impart unwanted odour, colour, or flavour).

Analysis. Determination of the elemental composition of isoprenoids seldom presents difficulty because many of them are hydrocarbons, and a simple, reliable procedure for quantitative analysis of carbon and hydrogen has been available since the early 19th century. The only other element commonly present in isoprenoids is oxygen, which does not interfere with the analysis for carbon and hydrogen, although it is difficult to determine directly: usually it is assumed to comprise the proportion of the compound not accounted for as carbon, hydrogen, and any other elements that have been measured.

Classical structural determination

Determination of structure. Assignment of structures to isoprenoids presented a very difficult problem. In the period when the monoterpenes were first investigated, the method of determining the structure of an organic compound was based entirely on studies of chemical reactions, usually those in which one or a few bonds in the compound were broken, giving successively simpler compounds, each of which was isolated, purified, and analyzed. These sequences of reactions eventually led to compounds possessing known structures, and the path back to the original substance was inferred from knowledge of the structural changes associated with the reactions employed. Frequently it was impossible to distinguish among alternate proposals; that is, more than one structure would be considered consistent with the information available. The situation was aggravated by the facts that many of the isoprenoids possess structures of a kind never previously observed, in which carbon atoms are linked together in small rings, and that, in many isoprenoids, the original carbon skeleton changes to a different one in the course of a reaction. One of the most difficult cases was the compound camphor, for which more than 30 different structures were proposed before the correct one (**14**) was established.



14

Modern techniques for determination of molecular structure lean heavily upon correlations between structural features and physical properties; many relationships of this kind have been established. Measurements of the wavelengths and intensities of absorption of ultraviolet, visible, and infrared light allow the chemist to draw very reliable inferences concerning the presence or absence of certain atomic groups and the distribution of electrons in the chemical bonding system. Determination of the intensity of magnetic fields at the nuclei of the atoms, especially the hydrogen atoms, permits him to deduce the molecular environment of those atoms. For example, the presence of an oxygen atom in the molecule of camphor was evident from the elemental analysis, which was performed in 1833, but the fact that that atom is part of a carbonyl ($>C{=}O$) group was established only in 1883, after many unsuccessful attempts to bring about a chemical reaction that would reveal that group. Nowadays, the presence of a carbonyl group would be instantly recognized by its characteristic absorption of infrared light. In favourable cases, complete structures have been revealed by X-ray crystal analysis. Mass spectra and optical rotatory dispersion measurements often elucidate fine structural details.

Laboratory synthesis. Once a structural formula has been deduced for a compound, the proposed structure becomes the objective of chemical synthesis as a test of the validity of the assignment. In certain cases, synthesis has been undertaken when the available evidence is consistent with more than one structure, but it appears simpler to synthesize all the likely candidates and to compare each of them with the unidentified compound than to seek further information from studies of chemical or physical properties.

In selecting a sequence of reactions by which a compound is to be synthesized for the purpose of confirming the assignment of a proposed structure, great care must be taken that, at every stage along the way, the transformations have occurred as predicted by the scheme. Such assurance is ordinarily provided by limiting the choice of reactions to those that have been thoroughly studied and shown to proceed in predictable ways and by careful examination of each intermediate compound to verify its identity. The first laboratory synthesis of camphor required 11 reactions and was a major achievement at the time (1903), although much more complicated compounds have been synthesized since then.

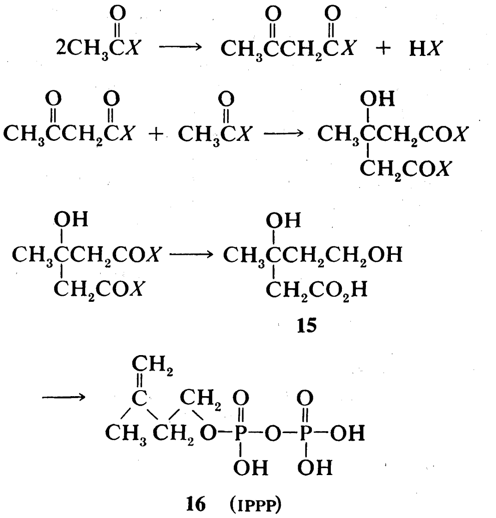
Camphor has been synthesized commercially by a series of reactions utilizing isoprenoids obtained from turpentine as the starting materials. Vitamin A is another isoprenoid manufactured by synthetic processes; several alternative reaction sequences have been developed for this compound.

Isoprenoids from acetate

Biosynthesis. By the late 1940s, the ubiquity of the five-carbon isoprene unit had been recognized for a long time, but the identity of the compounds actually involved in the physiological assembly of the isoprenoids was not known, although it had been suggested that they were somehow built up from acetic acid, the molecule of which contains only two carbon atoms. During the 1950s, it was shown that the synthesis of isoprenoids in nature indeed begins with acetylcoenzyme A, sometimes called activated acetate, a compound derived from acetic acid and coenzyme A, a complex substance that participates in many reactions that are controlled by enzymes. Previously unknown compounds, mevalonic acid and isopentenyl pyrophosphate (IPPP), occur as important intermediates in the process. The 1964 Nobel Prize for Physiology or Medicine was awarded to two German-born biochemists, Konrad Bloch and Feodor Lynen, for their contributions to this research.

The molecule of isopentenyl pyrophosphate is the long-sought physiological isoprene unit, containing five carbon atoms bonded together in the pattern discerned by Wallach in the 19th century. Most of the natural isoprenoids have structures that reveal the joining of two, three, four, six, or eight isoprene units; rubber is an isoprenoid in which several thousands of these units are connected in a long chain.

The process by which acetic acid units combine to form mevalonic acid (**15**) and isopentenyl pyrophosphate (**16**) is indicated by the equations below, in which *X* represents the complicated molecular structure of coenzyme A.



$$2CH_3\overset{O}{\overset{\|}{C}}X \longrightarrow CH_3\overset{O}{\overset{\|}{C}}CH_2\overset{O}{\overset{\|}{C}}X + HX$$

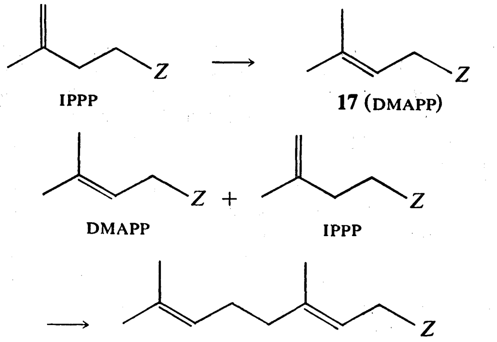
$$CH_3\overset{O}{\overset{\|}{C}}CH_2\overset{O}{\overset{\|}{C}}X + CH_3\overset{O}{\overset{\|}{C}}X \longrightarrow CH_3\overset{OH}{\overset{|}{\underset{\underset{CH_2COX}{|}}{C}}}CH_2COX$$

$$CH_3\overset{OH}{\overset{|}{\underset{\underset{CH_2COX}{|}}{C}}}CH_2COX \longrightarrow CH_3\overset{OH}{\overset{|}{\underset{\underset{CH_2CO_2H}{|}}{C}}}CH_2CH_2OH$$

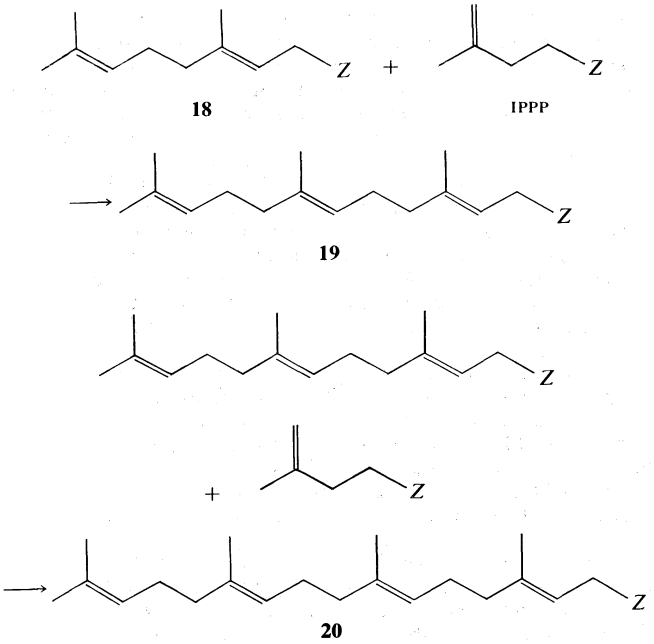
15

16 (IPPP)

The formation of geranyl pyrophosphate (**18**), the precursor of the monoterpenes, from two molecules of IPPP requires that one of them be transformed to dimethylallyl pyrophosphate (DMAPP, **17**). In the equations below, only the covalent bonds of the carbon skeletons are shown, and Z stands for the pyrophosphate group.



A similar reaction of geranyl pyrophosphate with IPPP leads to the 15-carbon compound, farnesyl pyrophosphate (**19**), from which the sesquiterpenes are derived, which, in turn, is converted to the 20-carbon precursor (**20**) of the diterpenes.



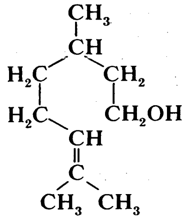
All these reactions produce substances in which the isoprene units are joined head-to-tail; many of the larger isoprenoid molecules (triterpenes and tetraterpenes) are not built up by further incorporation of IPPP units but by tail-to-tail coupling of 15- or 20-carbon compounds. The structure of squalene (**13**), a triterpene, illustrates this point. The tetraterpene carotenoids apparently arise from a similar reaction of two 20-carbon isoprenoids.

COMPARATIVE SURVEY OF ISOPRENOID COMPOUNDS

Monoterpenes. The monoterpenes are isolated from their natural sources by distillation of the plant matter with steam. They are volatile oils, less dense than water, and have normal boiling points in the range 150°–185° C (300°–365° F). Purification is usually achieved by fractional distillation at reduced pressures or by regeneration from a crystalline derivative. Acyclic monoterpene hydrocarbons are few in number, but their oxygenated derivatives are more widespread in nature and of greater importance.

Acyclic monoterpenes

Important oxygenated acyclic monoterpene derivatives include the terpene alcohol citronellol (**21**) and the corresponding aldehyde citronellal (**22**), both of which occur in oil of citronella, as well as citral (**23**), found in lemongrass oil, and geraniol (**24**), which occurs in Turkish geranium oil.



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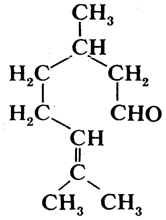

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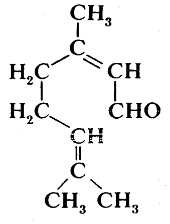

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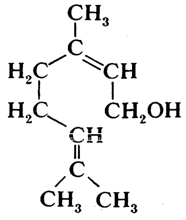

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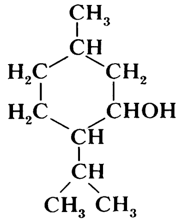
Citronellal is converted by treatment with acid into the monocyclic monoterpene alcohol isopulegol, from which a mixture of stereoisomeric menthols (**25**) is produced on catalytic hydrogenation. The process is used commercially to supplement the natural sources of menthol (oil of peppermint), widely used as a flavouring and in medicinal preparations. Citral (**23**), upon reduction with sodium amalgam, yields geraniol (**24**), an important component of rose perfumes. Citral may be condensed with acetone to yield the important intermediate pseudoionone (**26**),



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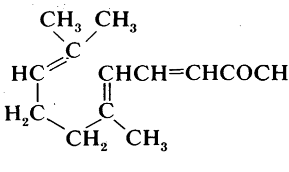

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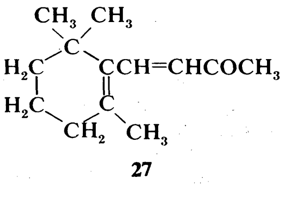
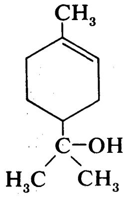


27

from which *β*-ionone (**27**) results on treatment with acid. Although *β*-ionone cannot be regarded as a terpene, it is of great importance as a starting material for the synthesis of vitamin A (**12**) and as a component of violet-scented perfumes.

Monocyclic monoterpenes

Limonene (**10**), an oil of normal boiling point 178° C (352° F), is a major component of orange and lemon oils and is typical of the monocyclic monoterpene hydrocarbons. Others of this class are terpinolene, *α*- and *β*-phellandrene, and *α*-, *β*-, and *γ*-terpinene, all of which have the same carbon skeleton as limonene and differ only in the location of the two carbon-to-carbon double bonds. Limonene is optically active (it rotates the plane of polarized light), as are most of the terpenes and their derivatives that contain an asymmetric carbon atom—that is, one bonded to four different groups. Limonene is converted to isoprene by contact with a heated metallic filament. Few commercial uses, other than as flavourings, exist for the monocyclic monoterpene hydrocarbons. Menthol (**25**), which has already been mentioned, and the oxygenated derivatives *α*-terpineol (**28**) and terpin (terpin hydrate; **29**) are chemicals of commerce. Mixtures of terpin, *α*-terpineol, terpinolene, and the terpinenes result from the treatment of *α*-pinene (**11**) with acid, and the mixture finds use as pine oil, an inexpensive disinfectant, deodorant, and wetting agent.



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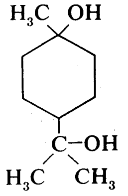

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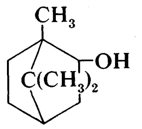
Bicyclic monoterpenes

Of normal boiling point 156° C (313° F), *α*-pinene is representative of the bicyclic monoterpenes and is the most abundant and important monoterpene. It is the major component of ordinary turpentine, which is prepared from pine trees or stumps either by extraction followed by rectification or by distillation with steam. It is a major component of sulfate turpentine, a by-product of the manufacture of paper, and is important as a component of paints and varnishes and as a raw material for the production of a wide variety of products employed in chemical industry.

Its use in coating materials depends upon its good properties as a solvent and upon its conversion by air oxidation into a polymeric, resinous film. Its function as a starting material for conversion to more useful products results from its abundance and its structure, which has a high degree of chemical reactivity. The chemical reactions of *α*-pinene have been studied more thoroughly than those of any other terpene, and the study has contributed greatly to the understanding of molecular rearrangements in organic chemistry.

Treatment of *α*-pinene with acids under various conditions leads to a host of products, among which are terpinolene, the terpinenes, *α*-terpineol (**28**), and terpin (**29**), previously mentioned, and, in addition, borneol (**30**), fenchyl alcohol (**31**), and the hydrocarbon camphene (**32**).



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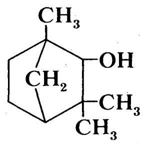

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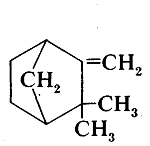

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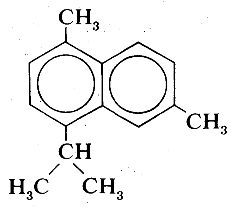
The formation of the latter three compounds involves molecular rearrangement, and advantage has been taken of the structural changes to provide a commercial synthesis of the important bicyclic terpene ketone, camphor (**14**).

Sesquiterpenes. Sesquiterpenes, $C_{15}H_{24}$, are isolated from their natural sources by distillation with steam or by extraction because they are of lower volatility than the monoterpenes. They are purified by fractional distillation *in vacuo* or by chromatography. The sesquiterpenes demonstrate an even greater complexity of structure than the monoterpenes, and oxygenated sesquiterpenes are commonly encountered.



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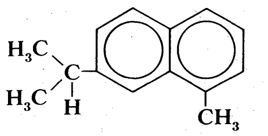

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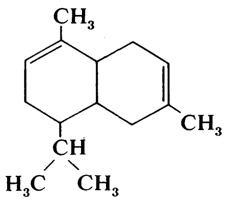
Two arrangements of isoprene units are found in bicyclic sesquiterpenes, the cadalene (**33**) and the eudalene (**34**) types, and the carbon skeleton of a sesquiterpene may frequently be determined by heating it with sulfur or selenium to effect dehydrogenation to the corresponding naphthalenic hydrocarbons: cadalene, 4-isopropyl-1,6-dimethylnaphthalene; or eudalene, 7-isopropyl-1-methylnaphthalene. In those cases in which sulfur dehydrogenation fails to yield information about the carbon skeleton of a sesquiterpene, a systematic degradation by oxidation to compounds of known structure is necessary.



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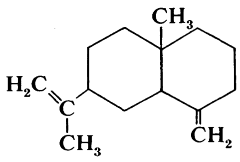

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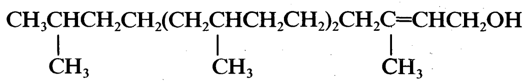
Cadinene (**35**), the principal component of oils of cubeb and cade, is a typical sesquiterpene of the cadalene type. It is an optically active oil of normal boiling point 274° C (525° F). *β*-Selinene (**36**), present in celery oil, is typical of the eudalene type.

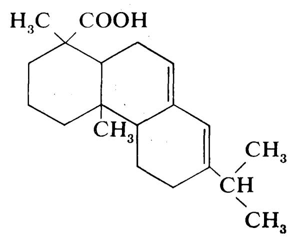
Diterpenes. Phytol (**37**), an oxygenated acyclic diterpene, is an important building block of the chlorophyll molecule, from which it is obtained on treatment with alkali solution. The arrangement of isoprene units in phytol is identical with that in vitamin A (**12**), a monocyclic diterpene derivative, and is typical of the head-to-tail arrangement of isoprene units found in most terpenes.



$$\mathrm{\underset{CH_3}{CH_3CH}CH_2CH_2(CH_2\underset{CH_3}{CH}CH_2CH_2)_2CH_2\underset{CH_3}{C}{=}CHCH_2OH}$$

37

The commercial importance of the bicyclic monoterpene *α*-pinene is paralleled in the diterpenes by abietic acid (**38**), a tricyclic carboxylic acid that constitutes the major portion of rosin. Rosin is the nonvolatile portion of the oleoresin of members of the pine family and is the residue left after the isolation of turpentine. Rosin is used as such in the production of varnish and coating materials and, in the form of its sodium salt, for sizing paper, as an emulsifying agent in producing synthetic rubber, and in yellow laundry soap. It is among the cheapest organic acids.



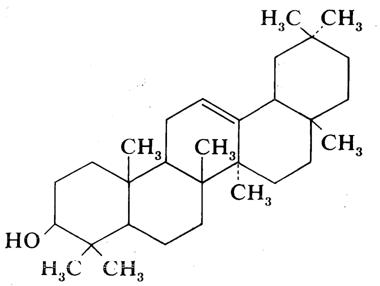
38

Squalene

Triterpenes. The acyclic triterpene hydrocarbon squalene (**13**) constitutes more than half of the liver oil of certain species of sharks and is otherwise rather widely distributed in nature. It has been found in other fish-liver oils, in vegetable oils, in fungi, and in human earwax and sebaceous secretions. The biochemical importance of squalene as a metabolic intermediate in the biosynthesis of cholesterol was demonstrated by the use of radioactive carbon labelling.

Although cholesterol is not a terpene, the demonstration that it has a terpene as precursor in metabolism represented a major advance in understanding the biochemical relationship between the two important classes of compounds.

Although tricyclic and tetracyclic triterpenes are known, by far the most abundant triterpenes found in nature are those having five carbon rings. The pentacyclic triterpenes, either free or combined with sugars in glycosides (saponins), occur in all parts of many plants. The structures of many of the pentacyclic triterpenes are now known in full detail; that of β-amyrin (**39**) exemplifies the important structural features of this class of substances. The best source of β-amyrin is the resin of the tropical tree elemi, although it has been isolated from more than 50 plant sources. The carbon skeleton of β-amyrin bears a striking relationship to those of squalene and cholesterol, and it has been shown that squalene is a common precursor of the pentacyclic triterpenes and the sterols in biosynthesis.



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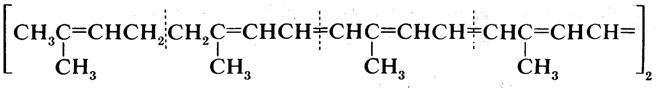
Carotenoids

Tetraterpenes. The large class of yellow, orange, or red, fat-soluble plant and animal pigments known as carotenoids are classed as tetraterpenes, although they have in general the molecular formula $C_{40}H_{56}$ rather than $C_{40}H_{64}$ as required by $(C_5H_8)_8$. The fact that their structures can be built up from isoprene units justifies their classification as terpenes. The carotenoids are isolated from their natural sources by solvent extraction and purified by chromatography.

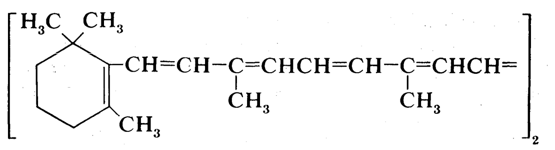
Lycopene (**40**) the red pigment of the ripe tomato, exemplifies the class of acyclic tetraterpenes. The dotted lines in the formula show the division into isoprene units, and it is to be noted that the usual head-to-tail attachment of isoprene units is interrupted in the centre of the molecule with a single tail-to-tail attachment that produces a symmetrical structure. This feature is generally encountered in the tetraterpenes, as is the long series of alternating single and double, carbon-to-carbon bonds (conjugated system) that is responsible for the absorption of light and hence the bright colours of the compounds.



$$\left[\mathrm{CH_3\underset{CH_3}{C}{=}CHCH_2\vdots CH_2\underset{CH_3}{C}{=}CHCH\vdots CH\underset{CH_3}{C}{=}CHCH\vdots CH\underset{CH_3}{C}{=}CHCH{=}}\right]_2$$

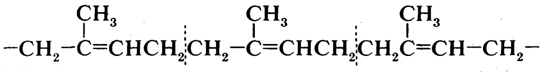
40

The most important and abundant tetraterpene is β-carotene (**41**), the principal yellow pigment of the carrot; β-carotene is of nutritional importance because the animal organism is apparently able to cleave the molecule at the point of symmetry with the production of vitamin A. The role of vitamin A and structurally related terpenoid molecules in the synthesis of the pigments in the eye necessary for vision has been demonstrated.



41

Polyterpenes. Rubber, which occurs in the latex of the rubber tree, is a polyterpene hydrocarbon, $(C_5H_8)_n$, in which n is 4,000–5,000. Chemical degradation by oxidation and X-ray-diffraction studies have revealed a repeating unit (**42**) in rubber. Division into isoprene units



$$\mathrm{-CH_2-\overset{CH_3}{C}{=}CHCH_2\vdots CH_2-\overset{CH_3}{C}{=}CHCH_2\vdots CH_2\overset{CH_3}{C}{=}CH-CH_2-}$$

42

is indicated. The vulcanization of rubber involves the establishment of cross-linking between the chains through sulfur atoms. Gutta-percha differs from rubber in the way in which methylene ($-CH_2-$) groups are arranged; in gutta-percha they are on opposite sides (*trans* arrangement) of the double bond, and in rubber they are on the same side (*cis* arrangement). (R.H.E./Ed.)

Steroids

Steroids are a class of naturally occurring organic substances and their derivatives that are of great importance in biology, medicine, and chemistry. The steroid group includes all sex hormones, adrenal cortical hormones, bile acids, and sterols of vertebrates, the molting hormones of insects, and many other physiologically active substances of animals and plants. Among the synthetic steroids of therapeutic value are a large number of anti-inflammatory agents, anabolic (growth-stimulating) agents, and oral contraceptives. Different categories of steroids are frequently distinguished from each other by names that relate to their biological source—*e.g.,* phytosterols (found in plants), adrenal steroids, bile acids—or to some important physiological function—*e.g.,* gestogens (promoting gestation), androgens (favouring development of masculine characteristics), cardiotonic steroids (facilitating proper heart function). Because steroids of similar function and origin tend to fall into families of closely related structure, biological categorization underlies an internationally accepted systematic chemical nomenclature.

HISTORICAL BACKGROUND

Studies of steroids commenced in the early 19th century with investigations of the unsaponifiable (*i.e.,* remaining undissolved after heating with excess of alkali) material, largely cholesterol, of animal fat and gallstones and of acids obtainable from bile. This early work, with which many of the noted chemists of the time were associated, led to the isolation of cholesterol and some bile acids in reasonable purity and established some significant features of their chemistry. Insight into the complex polycyclic steroid structure, however, came only after the turn of the 20th century, following the consolidation of chemical theory and the development of chemical techniques by which such molecules could be broken down step by step. Arduous studies, notably by the research groups of the German chemists Adolf Windaus and Heinrich Wieland, ultimately established the structures of cholesterol; of the related sterols, stigmasterol and ergosterol; and of the bile acids. Investigation of ergosterol was stimulated by the realization that it can be converted into vitamin D. Only in the final stages of this work (1932) was the arrangement of the component rings of the nucleus clarified by results obtained by pyrolytic (heat-induced bond-breaking) dehydrogenation and X-ray crystallography.

The foundations of steroid chemistry

With the foundations of steroid chemistry firmly laid, the next decade saw the elucidation of the structures of most of the physiologically potent steroid hormones of the gonads and the adrenal cortex. Added impetus was given to steroid research when the Americans Philip S.

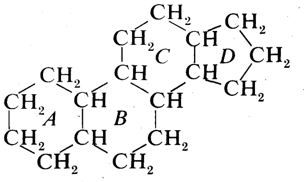
Hench and Edward C. Kendall announced in 1949 that the thitherto intractable symptoms of rheumatoid arthritis were dramatically alleviated by the adrenal hormone cortisone. New routes of synthesis of steroids were developed, and many novel analogues were therapeutically tested in a variety of disease states. From these beginnings there has developed a flourishing steroid pharmaceutical industry and, with it, a vastly expanded fundamental knowledge of steroid reactions that has influenced many other areas of chemistry.

Knowledge of the biochemistry of steroids has grown at a comparable rate, assisted by the use of radioisotopes and new analytical techniques. The metabolic pathways (sequences of chemical transformations in the body), both of synthesis and of decomposition, have become known in considerable detail for most steroids present in mammals, and much current research relates to control of these pathways and to the mechanisms by which steroid hormones exert their effects. The hormonal role of steroids in lower organisms is also of growing interest.

CHEMISTRY OF STEROIDS

Molecular structure

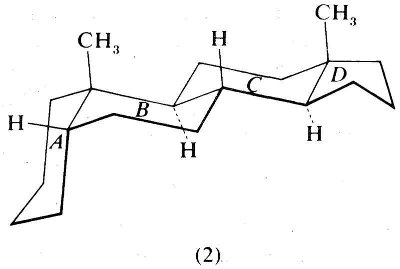
Steroid numbering system and nomenclature. All steroids are related to a characteristic molecular structure composed of 17 carbon atoms—arranged in four rings conventionally denoted by the letters *A, B, C,* and *D*—bonded to 28 hydrogen atoms. This parent structure, indicated in Diagram 1,



(1)

named gonane and often referred to as the steroid nucleus, may be modified in a practically unlimited number of ways by removal, replacement, or addition of a few atoms at a time; hundreds of steroids have been isolated from plants and animals, and thousands more have been prepared by chemical treatment of natural steroids or by synthesis from simpler compounds.

The steroid nucleus is a three-dimensional structure, and atoms or groups are attached to it by spatially directed bonds. Although many stereoisomers (compounds that are alike bond for bond but different in three-dimensional arrangement of atoms) of this nucleus are possible (and may be synthesized), the saturated nuclear structures of most classes of natural steroids are alike, except at the junction of rings *A* and *B*. Simplified three-dimensional diagrams may be used to illustrate stereochemical details. For example, Diagrams 2 and 3 show the alternative forms of the skeletal structure, androstane, common to a number of natural and synthetic steroids, in which the *A/B* ring fusions are called *cis-* and *trans-,* respectively. Note that at the common side of rings *A* and *B* in Diagram 2, bonds to the methyl group, CH_3, and to the hydrogen atom, H, both project upward from the general plane defined by the rest of the molecule, whereas in diagram 3, CH_3 projects up, and H projects down. Usually, however, steroid structures are represented as plane projection diagrams such as 4 and 5, which correspond to 2 and 3, respectively.



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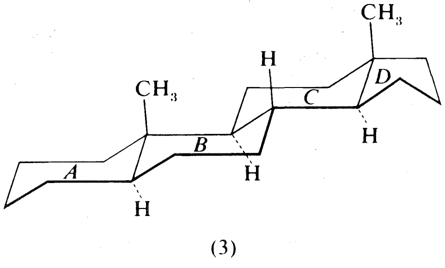

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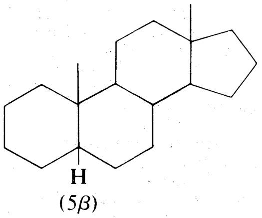

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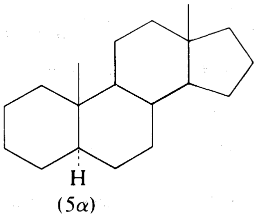
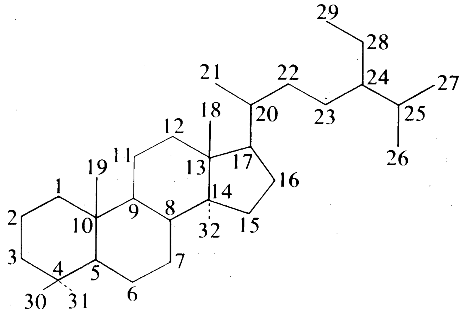


(5)

The stereochemistry of rings *A* and *B* must be specified by showing the orientation of the hydrogen atom attached at C-5 (that is, carbon atom number five; steroid numbering is explained below) as either above the plane of the diagram (designated *β*) or below it (*α*). The *α*-, *β*-symbolism is used similarly to indicate the orientation of any substituent group that is attached to a saturated (fully substituted) carbon of the steroid ring system. Groups attached to unsaturated carbons lie in the same plane as the adjacent carbons of the ring system (as in ethylene), and no orientation need be specified. When the orientation of a substituent is unknown, it is assigned the symbol *ξ*- (xi). Bonding of *β*-attached substituents is shown diagrammatically as in 4 by a full line; that of *α*-substituents by a broken line, as in 5, and that of *ξ*-substituents by a wavy line.

Each carbon atom of a steroid molecule is numbered, and the number is reserved to a particular position in the hypothetical parent skeletal structure (6) whether this position is occupied by a carbon atom or not.



(6)

Principal structural types

Steroids are named by modification of the names of skeletal root structures according to systematic rules agreed upon by the International Union of Pure and Applied Chemistry (IUPAC). By attaching prefixes and suffixes to the name of the appropriate root structure, the character of substituent groups or other structural modification is indicated. The prefixes and suffixes include numbers, called locants, indicative of the position in the carbon skeleton at which the modification occurs, and, where necessary, the orientation of a substituent is shown as *α*- or *β*-. The carbon atom at position 3, for example, is referred to as C-3; a hydroxyl group attached to C-3 is referred to as a 3-OH group or, more specifically, as a 3*α*-OH or 3*β*-OH group. In addition to differences in details of the steroid nucleus, the various classes of steroids are distinguished by variations in the size and structure of an atomic group (the side chain) attached at position 17. The derivations of the names of the more common root structures from those of naturally occurring compounds or classes of compounds of which they are most typical are indicated in Table 46. For unambiguous use of the names given in Table 46,

Table 46: Names of Fundamental Structures Most Often Used in Steroid Nomenclature

	carbon atoms present (as numbered in Diagram 6)	naturally occurring general classes	examples shown in the text
Gonane	1–17	none	gonane* (1)
Estrane	1–18	estrogens	estradiol (17 f)
Androstane	1–19	androgens	androstane* (2–5); testosterone (17 c); androstanedione (17 c)
Pregnane	1–21	gestogens and adrenal steroids	progesterone (17 a); cortisol (17 b); aldosterone (17 d); batrachotoxin (14)
Cholane	1–24	bile acids	cholic acid (10 a); sodium glycocholate (10 b)
Cholestane	1–27	sterols	cholesterol (16 d); scymnol (11 a)
Ergostane	1–28	sterols	ergosterol (8); cyasterone (22)
Stigmastane	1–29	sterols	stigmasterol (7); antheridiol (12)
Lanostane	1–27, 30–32	trimethyl sterols	lanosterol (16 c); 22,25-oxido-holothurinogenin (13)
Cardanolide	1–23	cardiac glycosides	digitoxigenin (23)
Bufanolide	1–24	toad poisons	bufotoxin (24)
Spirostan	1–27	sapogenins	dioscin (25)

*Gonane and androstane themselves do not occur in nature.

the orientation (*α* or *β*) of hydrogen at C-5 must be specified. If no other modification is indicated, the nucleus is assumed to be as shown in 2 and 3, except in the cardanolides and bufanolides: compounds of these types characteristically possess the 5*β*,14*β* configurations which, however, are specified.

For brevity in discussion and in trivial nomenclature, a number of prefixes are often attached, with locants, to the names of steroids to indicate specific modifications of the structure. In addition to the usual chemical notations for substituent groups replacing hydrogen atoms (*e.g.*, methyl-, chloro-, hydroxy-, oxo-), the following prefixes are commonly used: dehydro- (lacking two hydrogen atoms from adjacent positions); dihydro- (possessing two additional hydrogen atoms in adjacent positions); deoxy- (hydroxyl group replaced by a hydrogen atom); epi- (differing in configuration of a carbon atom bonded to two other carbon atoms); iso- (differing in configuration of a carbon atom bonded to three other carbon atoms); nor- (lacking one carbon atom); homo- (possessing one additional carbon atom); cyclo- (with a bond between two carbons that are normally not united); and seco- (with a carbon–carbon bond of the nucleus broken).

Polarity

Depending upon the number and character of their functional groups, steroid molecules may show diverse reactivities. Moreover, the reactivity of a functional group varies according to its location within the molecule (*e.g.*, esters are formed readily by 3-OH groups but only with difficulty by the 11*β*-OH group). An important property of steroids is polarity; *i.e.*, their solubility in oxygen-containing solvents (water, alcohols) rather than hydrocarbon solvents (hexane, benzene). Hydroxyl, ketonic, or ionizable (*i.e.*, capable of dissociating to form electrically charged particles) groups in a steroid molecule increase its polarity to an extent that is strongly influenced by the spatial arrangement of the atoms within the molecule.

Methods of isolation. Procedures for isolation of steroids differ according to the chemical nature of the steroids and the scale and purpose of the isolation. Steroids are isolated from natural sources by extraction with organic solvents, in which they usually dissolve more readily than in the aqueous tissue fluids. The source material often is treated initially with an alcoholic solvent, which dehydrates it, denatures (renders insoluble) proteins associated with the steroids, and dissolves many steroids. Saponification (treatment with aqueous alkali) either of whole tissues or of substances extracted from them by alcohol splits the molecules of sterol esters, triglycerides, and other fatty esters and permits the extraction of the sterols by means of water-immiscible solvents, such as hexane or ether, with considerable purification. Intact sterol esters or hormonal steroids and their metabolites (compounds produced by biological transformation) that are sensitive to strong acids or alkalies, however, require essentially neutral conditions for isolation, and, although some procedures for analysis of urinary steroids employ acid treatment, milder hydrolysis, as by enzymes, is preferred. The acidity of some steroids allows them to be held in alkaline solution while nonacidic impurities are extracted with organic solvents.

Commercially, abundant steroids usually are purified by repeated crystallization from solvents. Small-scale laboratory isolations for investigative or assay purposes usually exploit differing polarities of the steroid and of its impurities, which may be separated by partition between solvents differing in polarity or by chromatography (see below). Occasionally, special reagents may selectively precipitate or otherwise sequester the desired steroid. A classical example is the precipitation of 3*β*-hydroxy sterols such as cholesterol by the natural steroid derivative digitonin. New steroids of great physiological interest often are isolated from tissue only with extreme difficulty because they are usually trace constituents. For example, 500 kilograms (1,100 pounds) of silkworm pupae yielded 25 milligrams (0.0008 ounce) of pure molting hormone, the steroid ecdysone (*i.e.*, 20×10^6-fold purification). In such cases each isolation step is followed by an assay for the relevant physiological activity to ensure that the desired material is being purified. The percentage recovery of known steroid hormones during their assay in small biological samples usually is assessed by adding a trace of the same steroid in radioactive form to the initial sample, followed by radioassay (analysis based on radioactivity) after purification is complete. The efficiency of recovery of the radioactive steroid is assumed to be the same as that of the natural substance.

Determination of structure and methods of analysis. The systematic, stepwise breakdown by chemical methods of the steroid ring systems, used in early investigations of structure, is now mainly of historical interest. The small number of different nuclear structures found in steroids often has permitted establishment of the structure of a new steroid by conversion to related compounds of known structure. Structure elucidation in the steroid field, as in all areas of organic chemistry, now depends heavily upon physical methods, particularly nuclear magnetic resonance, infrared spectroscopy, optical rotatory dispersion, mass spectrometry, and X-ray crystallography. Data obtained by these methods reinforce and often replace the classical criteria of characterization of steroids: melting point, optical rotation, elemental analysis, and ultraviolet absorption at a fixed wavelength.

Chromatography

Chromatography is a crucial technique in steroid chemistry. The behaviour of a steroid in selected chromatographic systems often identifies it with a high degree of probability. The identification may be made virtually certain by the conversion of the material to derivatives that in turn are examined chromatographically. Abundant data for the behaviour of steroids in paper chromatography, thin-layer chromatography, and gas–liquid chromatography show that individual features of molecular structure determine the chromatographic properties of steroids in a predictable manner. Because gas-liquid chromatography is the most precisely controllable technique, it provides the most valuable data for structural identification. The gas–liquid chromatograph linked directly to the mass spectrometer permits characteristic mass-spectral fragmentation patterns and critical gas–liquid chromatographic data to be obtained simultaneously, using a sample containing less than a microgram of a steroid. This powerful technique is of growing importance in the structural analysis of steroids in extracts of such body fluids as blood and urine.

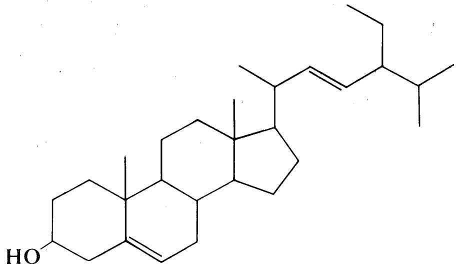
Synthesis of steroids. *Total synthesis.* The total synthesis of any steroid (*i.e.*, synthesis from simple organic chemicals by reactions that follow known pathways) is formally the final proof of its chemical structure. To reproduce all stereochemical details of the naturally occurring material, however, is an exacting intellectual and technical challenge that has been met with extraordinary ingenuity and originality by several chemists.

In most total syntheses of steroids, a monocyclic starting material such as a quinone provides one ring upon which the other rings of the nucleus are elaborated, step by step, by condensation reactions with smaller molecules to give the desired stereochemistry in successive ring fusions. Each new ring closure must also provide functional groups that can be used in building up the next ring. In a quite different approach, stereochemical control of ring fusions is achieved by using the fact that, under acidic conditions, open-chain molecules containing suitably located double bonds cyclize to multi-ring structures that have the necessary stereochemistry and that can be relatively easily converted to steroids. From its analogy with the cyclization of squalene 2,3-oxide to lanosterol in the biosynthesis of cholesterol (see below), this method is said to involve "biogenetic-type cyclization."

Partial synthesis. Although total synthesis of steroids has proved commercially feasible, it is often more practical to prepare them by partial synthesis; that is, by modification of other steroids that are naturally abundant. To be useful as a starting material for partial synthesis, the naturally occurring steroid must possess a molecular structure that can be easily converted to that of the desired product. For the synthesis of cortisol, cortisone, and their analogues, which carry an oxygen function at C-11, a pre-existing oxygen function at this position or at the adjacent C-12 is highly desirable. Indeed, prior to the advent of methods for microbiological oxidation, this was a crucial requirement, since the introduction of any functional group at C-11 of most steroids was extremely difficult.

Commercial synthesis

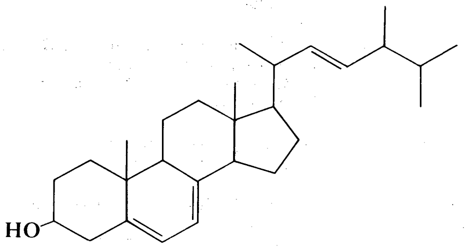
In the early commercial synthesis of androgenic steroids, cholesterol was the main starting material. Cholic acid and deoxycholic acid, inexpensive by-products from slaughterhouses, were starting materials for production of cortisone. The bulk of steroid drugs now is manufactured from the abundant steroids of plant origin, notably the sapogenins. Diosgenin, obtainable from several varieties of *Dioscorea,* a Mexican yam, is used in the commercial manufacture of progesterone, which can be converted to androgenic and estrogenic hormones and to the more complex adrenal steroid hormones, such as cortisone and cortisol. A most important advance in this field was the discovery that microorganisms such as *Rhizopus nigricans* introduce hydroxyl groups into a variety of steroids at C-11 and elsewhere: they are used in the commercial synthesis of a large number of steroid hormone analogues. A sapogenin, hecogenin, obtainable in quantity from the waste of sisal plants, is used for synthesis of cortisol. Stigmasterol, which is readily obtainable from soybean oil, can be transformed easily to progesterone and to other hormones, and commercial processes based on this sterol have been developed.

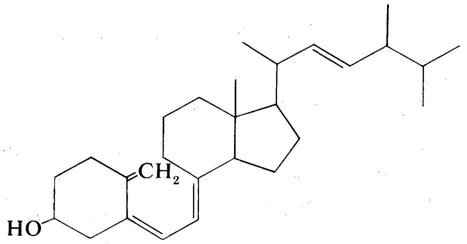

stigmasterol (7)

NATURAL DISTRIBUTION AND FUNCTIONS

Sterols and bile acids. The most generally abundant steroids are sterols, which occur in all tissues of animals, green plants, fungi, and yeasts. Evidence for the presence of steroids in bacteria and in primitive blue-green algae is conflicting. The major sterols (Table 47) of most tissues are accompanied by traces of their precursors—lanosterol in animals and cycloartenol in plants—and of intermediates between these compounds and their major sterol products. In mammalian skin, one precursor of cholesterol, 7-dehydrocholesterol, is converted by solar ultraviolet light to cholecalciferol, vitamin D_3, which controls calcification of bone by regulating intestinal absorption of calcium. The disease rickets, which results from lack of exposure to sunlight, can be treated by administration of the vitamin or of the corresponding derivative of ergosterol, ergocalciferol (vitamin D_2).

Rickets and atherosclerosis


ergosterol (8)


ergocalciferol (9)

Sterols are present in tissues in both the non-esterified (free) form and as esters of aliphatic fatty acids. In the disease atherosclerosis, fatty materials containing cholesterol form deposits (plaques), especially in the walls of the major blood vessels, and vascular function may be fatally impaired. The disease has many contributory factors but typically is associated with elevated concentrations of cholesterol in the blood plasma. One aim of medical treatment is to lower the plasma cholesterol level.

Free sterols appear to stabilize the structures of cellular and intracellular membranes. Because the sheath of nerve fibres is a deposit of many layers of the membranes of neighbouring cells, mature mammalian nerve tissue (*e.g.,* beef brain) is the richest source of cholesterol. Cholesterol also is converted in animals to steroids that have a variety of essential functions, and in plants to steroids the functions of which are less clearly understood. The bile acids

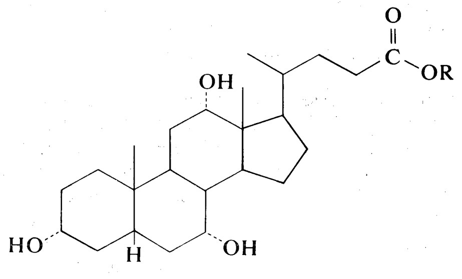
Table 47: Some Important Sterols and Their Sources

common name	systematic name	occurrence
Cholesterol	5-cholesten-3β-ol	principal sterol of most animals and all vertebrate tissues
Coprostanol	5β-cholestan-3α-ol	feces of vertebrates
Cholestanol	5α-cholestan-3β-ol	minor vertebrate sterol: guinea pig and rabbit adrenal
Lathosterol	5α-cholest-7-en-3β-ol	vertebrate skin, intestine
7-Dehydrocholesterol	5,7-cholestadien-3β-ol	mammalian skin, intestine
Desmosterol	5,24-cholestadien-3β-ol	chick embryo, barnacle (*Balanus glandula*)
Zymosterol	5α-cholesta-8,24-dien-3β-ol	minor sterol of yeasts
Ergosterol	5,7,22-ergostatrien-3β-ol	principal sterol of yeasts, ergot (*Claviceps purpurea*), and other fungi
Stigmasterol	5,22-stigmastadien-3β-ol	most green plants, soy beans
Sitosterol	5-stigmasten-3β-ol	most green plants, wheat germ
Fucosterol	5,24(28)-stigmastadien-3β-ol	principal sterol of marine brown algae (*Fucus* species)
Lanosterol	8,24-lanostadien-3β-ol	skin, sheep wool fat, yeasts
Lophenol	4α-methyl-5α-cholest-7-en-3β-ol	skin, intestine, feces, cactus (*Lophocerus schotti*)
Cycloartenol	9,19-cyclo-24-lanosten-3β-ol	generally minor sterol of green plants (*Artocarpus* species)

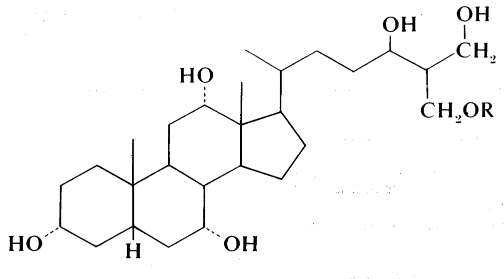
Table 48: Evolutionary Relationships of Bile Alcohols and Bile Acids

species	bile alcohol or bile acid (trivial name)
Elasmobranch fish	
Myxine glutinosa (hagfish)	5α-cholestane-3β,7α,16α,26-tetrol (myxinol)
Mustelus manazo (dogfish)	5β-cholestane-3α,7α,12α,24,26,27-hexol (scymnol)
Teleostian fish	
Latimeria chalumnae (coelacanth)	5α-cholestane-3β,7α,12α,26,27-pentol (latimerol)
Cyprinus carpio (carp)	5α-cholestane-3α,7α,12α,26,27-pentol (5α-cyprinol)
Amphibians	
Rana temporaria (frog)	5α,27-norcholestane-3α,7α,12α,24,26-pentol (5α-ranol)
Bufo vulgaris japonica (toad)	5β-cholestane-3α,7α,12α,25,26-pentol (5β-bufol)
Reptiles	
Alligator mississippiensis (alligator)	3α,7α,12α-trihydroxy-5β-cholestan-27-oic acid (trihydroxycoprostanoic acid)
Bitis arietans (puff adder)	3α,12α,23-trihydroxy-5α-cholanoic acid (bitocholic acid)
Bird	
Anser domesticus (goose)	3α,7α-dihydroxy-5β-cholanoic acid (chenodeoxycholic acid)
Mammals	
Man and most (but not all) species	3α,7α,12α-trihydroxy-5β-cholanoic acid (cholic acid) 3α,12α-dihydroxy-5β-cholanoic acid (deoxycholic acid) 3α-hydroxy-5β-cholanoic acid (lithocholic acid)

(cholanoic acids, earlier called cholanic acids) of higher vertebrates form conjugates with the amino acids taurine and glycine, and the bile alcohols (cholane derivatives) of lower animals form esters with sulfuric acid (sulfates; Table 48). These conjugates and sulfates enter the intestine as sodium salts and assist in the emulsification and absorption of dietary fat, processes that may be impaired when bile acid secretion is reduced, as in some liver diseases and in obstructive jaundice. The mixture of bile acids found in feces reflects the actions of intestinal microorganisms upon the primary bile acid secretory products (*e.g.*, deoxycholic acid arises by bacterial transformation of cholic acid).



cholic acid (10a; R = H) and sodium glycocholate (10b; R = $NHCH_2COO^-Na^+$)



scymnol (11a; R = H) and sodium scymnol sulfate (11b; R = $SO_3^-Na^+$)

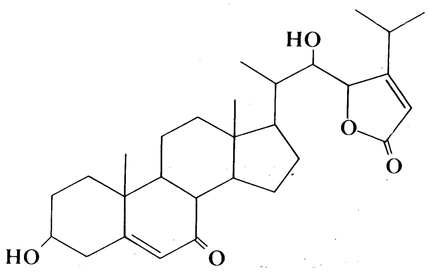
Sex hormones. Steroids that have a phenolic ring *A* (that is, those in which ring *A* is aromatic and bears a hydroxyl group) are ubiquitous products of the ovary of vertebrate animals. These are the estrogens, of which estradiol is the most potent. They maintain the female reproductive tissues in a fully functional condition, promote the estrous state of preparedness for mating, and stimulate development of the mammary glands and of other feminine characteristics. Estrogenic steroids have been isolated from urines of pregnant female mammals of many species including man, from placental and adrenal tissues, and, unexpectedly, from the testes and urines of stallions.

Estrogens and androgens

The corpus luteum, a modification of vertebrate ovarian tissue that forms following ovulation, produces the gestogens, which include progesterone and its derivatives. Gestogens also are secreted by the adrenals and placenta in some species. These compounds, in combination with the estrogens, regulate the metabolism of the uterus to permit implantation and subsequent development of the fertilized ovum in mammals. In birds, estrogens and gestogens stimulate the development of the oviduct and its secretion of albumin. Estrogens and gestogens suppress ovulation (release of the mature egg cell from the ovary); this fact is the basis of action of steroid anti-fertility drugs (see below). Estrogens and gestogens occur in primitive invertebrates, but their functions in those animals are obscure.

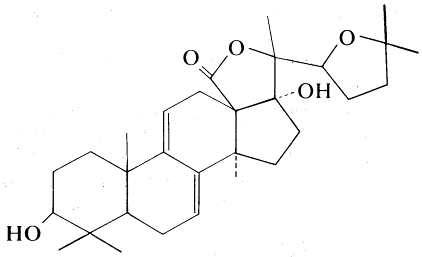
In male vertebrates, the androgens—steroids secreted by the testes—maintain spermatogenesis and the tissues of the reproductive tract. They promote male sexual behaviour and aggressiveness, muscular development, and, in man, the growth of facial and body hair and deepening of the voice. Testosterone and androstenedione are the principal androgens of the testes. Testosterone is more potent than androstenedione, but in the sexual tissues it appears to be converted to 5α-dihydrotestosterone, an even more potent androgen.

Adrenal hormones. The adrenal cortex of vertebrates synthesizes oxygenated progesterone derivatives. These compounds are hormones that are vital to survival and are classified according to their biological activity. The glucocorticoids promote deposition of glycogen in the liver and the breakdown of body proteins. Mineralocorticoids stimulate retention of sodium in the extracellular body fluids. Cortisol is the principal glucocorticoid in many species including man; in most rodents, this role is filled by corticosterone. The most potent mineralocorticoid of all species is aldosterone. Aldosterone has about 20 percent of the glucocorticoid activity of cortisol, which, conversely, has about 0.1 percent of the mineralocorticoid activity of aldosterone. Either steroid can maintain life in an animal from which the adrenal glands have been removed. The secretion of glucocorticoids is exquisitely responsive to injury and fear in animals and is primarily responsible for metabolic adaptation to stressful conditions. Failure of the adrenal cortex in man gives rise to Addison's disease, a formerly fatal condition that can now be successfully treated with synthetic adrenal steroids.



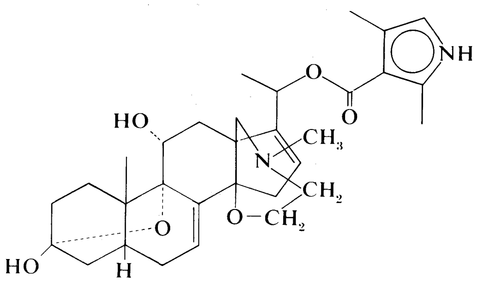
antheridiol (12)

Steroids of lower organisms. An area of increasing interest is the role of steroids in the reproduction, development, and self-defense of lower organisms. In an aquatic fungus, *Achlya bisexualis,* the steroid antheridiol (12) of the female stimulates male gamete formation. Insects and crustaceans produce the ecdysones, steroid hormones that promote molting and the development of adult characteristics (Table 49). Many plants, especially ferns and conifers, contain steroids that may protect them against some predatory insects, although this function is not established. Progesterone, 11-deoxycorticosterone, and related steroids with no known endocrine function in insects are released into the water by several species of water beetles to repel predatory fish, and the sea cucumbers (Holothuroideae) produce the holothurinogenins, a group of lanosterol derivatives toxic to nervous tissue.

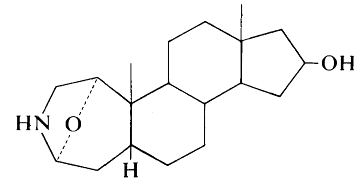


22,25-oxidoholothurinogenin (13)

Cardanolide and bufanolide derivatives, found in many plants and in the skin of toads, cause vomiting, visual disturbances, and slowing of the heart in vertebrates and are strong deterrents to predators. Birds and other predators avoid certain grasshoppers and butterflies that store cardenolides of the plants upon which they feed. The skin of the Colombian frog, *Phyllobates aurotaenia,* produces a deadly alkaloid, batrachotoxin, which is used by tribal peoples as an arrow poison. The skin of salamanders secretes a comparably poisonous alkaloid—samandarin.



batrachotoxin (14)



samandarin (15)

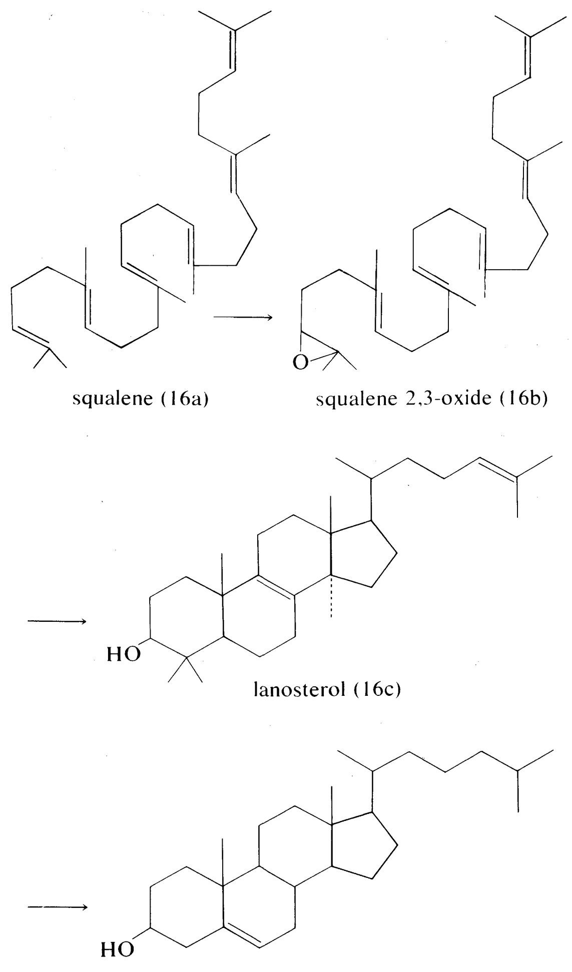
Many steroid alkaloids occur in plants, but their functions, like those of the steroid saponins, are unknown. It is possible that the taste of many of these compounds deters grazing animals or that they attract certain insect species to the plant.

BIOSYNTHESIS AND METABOLISM OF STERIODS

Cholesterol. Steroids are probably synthesized in all vertebrates and in many invertebrates by the same pathway, which includes cholesterol. Biosynthesis (formation in intact organisms) of cholesterol is especially vigorous in the liver of vertebrates but also occurs in the intestine, gonads, skin, and immature brain. It is barely detectable in the adult brain. The insects, certain mollusks, annelids, and some protozoa do not synthesize cholesterol but must obtain it, or a related sterol, in their diet.

Squalene and lanosterol

Cholesterol and other steroids are biosynthesized by extension of the enzyme pathway by which terpenoids are synthesized (see above *Isoprenoids*). Acetate fragments derived from common nutrient materials are converted into mevalonic acid, from which the terpenoid hydrocarbon squalene is formed. One end of the squalene molecule is then oxidized, giving squalene 2,3-oxide, which, by an intramolecular reaction (cyclization) and structural rearrangement, yields the steroid lanosterol. This enzyme-controlled reaction may be initiated by introduction of a positive charge into the oxide ring because it is remarkably similar to the nonenzymic, acid-catalyzed cyclizations of certain unsaturated hydrocarbons similar in structure to squalene. Cholesterol is formed from lanosterol by further structural changes.



squalene (16a) squalene 2,3-oxide (16b)

lanosterol (16c)

cholesterol (16d)

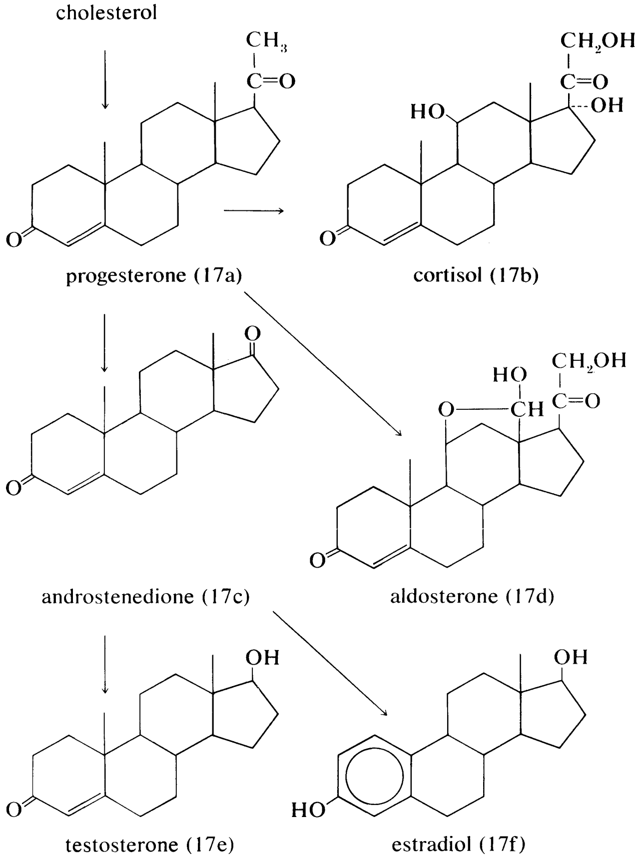
biogenesis of cholesterol

The principal forms in which cholesterol is excreted by vertebrates are the bile acids, which are synthesized in the liver. Their formation involves specific modifications of the steroid nucleus and formation of a carboxylic acid group that is linked to the amino acids taurine or glycine to give the forms in which the bile acids are secreted into the bile. The biosynthesis of cholesterol is influenced by feedback mechanisms that suppress the formation of mevalonic acid and, consequently, of cholesterol when levels of cholesterol and bile acids in the tissues are elevated.

Steroid hormones. In vertebrates, cholesterol is the central precursor of all steroid hormones secreted by the testes of the male, the ovaries of the female, and the adrenals of both sexes. These tissues share an embryonic tissue of origin and, in consequence, many enzymes for the transformation of cholesterol. A major (though not exclusive) common pathway involves conversion to progesterone. Progesterone is secreted by the corpus luteum of the ovary, but in the adrenal cortex it is further metabolized to steroid hormones (corticosteroids) such as cortisol and aldosterone. In both ovary and testis, progesterone is transformed further to the androgenic steroid androstenedione, which, together with its derivative testosterone, is secreted by the testis. In the ovary, androstenedione is modified to the estrogen estradiol. Normally, each organ secretes its own characteristic pattern of hormones, but in some disease states (*e.g.,* genetic defects and some tu-

Progesterone

mours of these endocrine glands) these patterns may be profoundly distorted.



biogenetic relationships of principal vertebrate steroid hormones

Many tissues, but mainly the liver, metabolize the steroid hormones to physiologically inactive products that are voided mainly in the urine, though some are also eliminated via the bile and, ultimately, the feces. Diagnosis of endocrine abnormalities may be assisted by analysis of urinary steroids. Urinary 17-ketosteroids (androstane derivatives with a >C=O function at C-17) arise principally through oxidation of adrenal steroid hormones in the liver and thus gauge secretion by the adrenal rather than by the testis. In pregnancy, the urinary excretion of pregnanediol, the principal metabolite of progesterone, measures placental progesterone output. Its decline before term may forewarn of abortion, which may be averted by administration of progestational hormones.

Steroid metabolism in plants. The early steps in the biosynthesis of steroids of both plants and animals are the same, except that in plants lanosterol is replaced by the related compound cycloartenol, which contains a three-membered ring (C-9, C-10, C-19) in lieu of the nuclear double bond of lanosterol. The side chains of the phytosterols, such as stigmasterol, and of the sterol ergosterol of yeasts and fungi contain extra carbon atoms that are incorporated in reactions involving *S*-adenosylmethionine, which donates methyl (CH_3) groups in numerous biological processes. Although most plant tissues contain only traces of cholesterol, this sterol is the biogenetic precursor of such important plant steroids as the sapogenins, cardiac glycosides, and steroid alkaloids. Because pregnane derivatives are intermediates in some of these transformations, plants and animals appear to have important features of steroid metabolism in common.

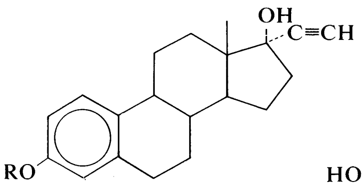
STRUCTURAL RELATIONSHIPS OF THE PRINCIPAL CATEGORIES OF STEROIDS

Sterols. A large group, the sterols, comprises the common 3-monohydroxy steroids of the cholestane, ergostane, and stigmastane series and their "methyl sterol" biogenetic precursors: lanosterol, cycloartenol, and certain derivatives of these sterols, such as lophenol (Table 47). Most sterols have a 3β-hydroxyl group, and many (though not the 4-methyl sterols) have a double bond between carbon atoms 5 and 6. Various sterols have double bonds at other positions in the nucleus corresponding to stages in the biosynthesis of cholesterol and other steroids that resemble it in the structure of ring *A*. Animal sterols, especially in embryonic tissues and skin, and phytosterols (*e.g.*, stigmasterol) also may have a double bond in the side chain. Sterols of feces (*e.g.*, coprostanol) have a 3α-hydroxyl group and *cis*- (5β) linked rings *A* and *B;* they are formed by metabolism of other sterols by intestinal bacteria. Certain sterols are transformed to calciferols (D vitamins) by ultraviolet light; this process occurs naturally in the skin and is used commercially in the manufacture of vitamin D_2 (ergocalciferol) from ergosterol and of vitamin D_3 (cholecalciferol) from synthetic 7-dehydrocholesterol.

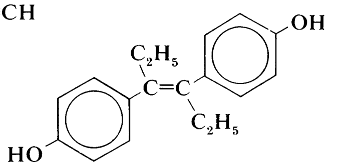
Bile acids and alcohols. The molecular structures of metabolites of cholesterol form an evolutionary series from the bile alcohols, such as myxinol and scymnol of the elasmobranch fishes and the related alcohols of some bony fishes and frogs, through the 5β-cholestanoic acids of crocodiles and alligators, to the 5β-cholanoic acids of the birds and mammals. This evolutionary trend is illustrated by the selected compounds in Table 48. They are not exclusively confined to the species indicated; for example, chenodeoxycholic acid is a major bile acid in man and many other mammals, and cholic acid is found in many nonmammalian species, together with primitive bile acids or alcohols that are not found in the mammals.

Estradiol

Estrogens. The estrogens of the ovary of vertebrates are steroids that are abundant in the urines of pregnant mares and of stallions. The most potent natural estrogen is estradiol; the less potent estrogens—estrone, estriol, and other oxygenated phenolic steroids—are metabolites of estradiol. Some species, notably the Equidae, secrete the less active estrogen equilenin. Estrone, synthesized from diosgenin, has been used as a starting material for synthesis of androgenic and progestational steroids lacking a C-19 methyl group (19-nor steroids). Synthetic estrogens, as estranol or mestranol, commonly used in oral contraceptives and for other therapeutic purposes, have acetylenic (containing triple bonds between carbon atoms) substituents. Nonsteroidal synthetic estrogens—*e.g.*, diethylstilbestrol and related compounds—are used clinically and also in animal husbandry to promote fattening of livestock and poultry and to improve the quality of their meat.



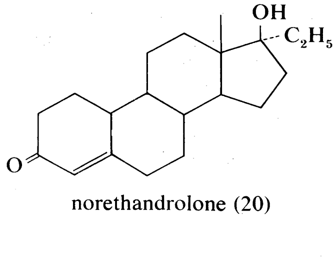
estranol (18a; R = H) and mestranol (18b; R = CH_3)



diethylstilbestrol (19)

Progestational steroids (gestogens, progestins). Three naturally occurring steroids of the corpus luteum and placenta have progestational action; these are progesterone and two of its metabolites. All possess an unsaturated ketonic structure in ring *A*. Pregnanediol, the main metabolite of progesterone, lacks both this structural feature and progestational activity.

Synthetic progestational steroids that are used in oral contraceptives and for other therapeutic purposes (see below) are derivatives of progesterone or of 19-nortestosterone. Among the latter is norethandrolone.



norethandrolone (20)

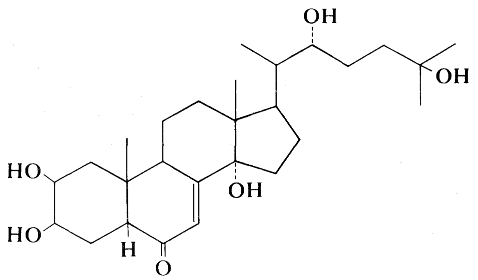
Table 49: Representative Ecdysones of Arthropods and Plants

compounds	variant of ecdysone structure	occurrence
C_{27} steroids		
Ecdysone (α-ecdysone) $2\beta,3\beta,14\alpha$,22,25-pentahydroxy -5β-cholest-7-en-6-one		silkworm, *Bombyx mori*; oak silk moth, *Antheraea pernyi*, and other insects; also the ferns *Pteridium aquilinum* and *Polypodium vulgare* and *Osmunda* species
Ecdysterone, crustecdysone (β-ecdysone)	20α-hydroxy-	*Bombyx mori*, *Antheraea pernyi* and other insects; crab, *Callinectes sapidus* and crayfish, *Jasus lalandei*; the conifer, *Podocarpus elatus*, the yew, *Taxus baccata* and several ferns
2-Deoxycrustecdysone	2-deoxy-20α-hydroxy-	*Jasus lalandei*
Callinecdysone A	20, 26-dihydroxy-, 25-deoxy-	*Callinectes sapidus* (pre-molting)
C_{28} steroids		
Callinecdysone B	20-hydroxy-, 24-methyl-	*Callinectes sapidus* (post-molting)
Makisterone B	20, 26-dihydroxy-, 24-methyl-25-deoxy-	*Podocarpus macrophyllus*
C_{29} steroids		
Makisterone C	20-hydroxy-, 24-ethyl-	*Podocarpus macrophyllus*
Cyasterone	20, 22-dihydroxystigmasteno-26, 28-lactone side chain	*Cyathula capitata* and several *Ajuga* species

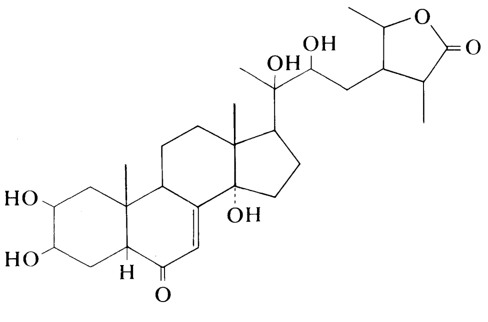
Androgens. Testosterone and androstenedione are the major testicular androgens. Several other less active androgens occur naturally. Major metabolites of testosterone are androsterone and etiocholanolone. The latter compound is androgenically inactive, but it is a pyrogen (*e.g.*, a fever-producing agent) that has been associated clinically with some febrile conditions. Various carboxylic acid esters and a number of synthetic analogues of testosterone are used therapeutically (see below).

Adrenal cortical hormones. The steroids of the adrenal cortex are progesterone derivatives that bear hydroxyl groups at positions 11, 17α, or 21. The potent mineralocortoid aldosterone carries an aldehyde function in place of the more usual C-18 methyl group. Glucocorticoid potency is higher in the trihydroxy derivative cortisol of man, monkeys, and dogs than in the dihydroxy steroid corticosterone of rats and mice. Every adrenal steroid hormone has a ketone group at C-3 and a double bond between C-4 and C-5. In the liver their physiological activities are lost when ring *A* is reduced, and they are partially inactivated when the 11β-hydroxyl is oxidized to a ketone group (as in cortisone and 11-dehydrocorticosterone). The adrenal corticoids are among the most chemically reactive of the steroid hormones; they are sensitive to strong acids and alkalies and to elevated temperatures.

In many synthetic analogues of cortisol and cortisone, physiological activity is modified. Derivatives with enhanced anti-inflammatory activity are especially valuable therapeutically and are discussed below.



α-ecdysone (21)



cyasterone (22), a phytoecdysone

Molting hormones

Ecdysones. The molting hormones (zooecdysones) of insects and crustaceans are generally derivatives of cholestane (Table 49). All possess a ketone group at position 6, a double bond between positions 7 and 8, and 2β, 3β, and 14α-hydroxyl groups. The side chain is hydroxylated at C-22 and variously at C-20, C-25, and C-26. Some of these compounds occur in plants, many of which also contain potent ecdysone analogues (phytoecdysones) with ergostane and stigmastane side chains.

The ecdysones show cross-species activity—that is, the ecdysone of one species induces molting in another species. The preponderance of ecdysones related to cholestane even in phytophagous insects probably reflects their capacity to convert dietary phytosterols to cholestane derivatives.

Cardiac glycosides and aglycones. Many species of plants contain toxic (specifically, heart-arresting) steroids of the cardanolide type as glycosides (compounds that contain structural groups derived from sugars) of up to four sugar residues, which may include glucose, rhamnose, and 10 other sugars peculiar to this group of natural products. Typically, these compounds (Table 50) are 5β-steroids and have 3β- and 14β-hydroxyl groups, but hydroxyl groups may occur in many other positions. In all cases, the aglycone (the steroid that results when the sugar groups are removed) is less active than its glycosides, but generally, activity declines with increasing numbers of sugar residues after the first. The structures of the sugars have important but not predictable effects on activity.

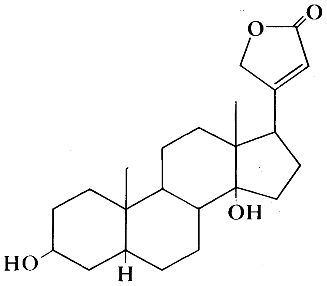
Table 50: Some Cardiac Glycosides and Their Cardenolide Aglycones

glycoside	aglycone	systematic name of aglycone
Digitoxin	digitoxigenin	$3\beta,14\beta$-dihydroxy-5β-card-20(22)-enolide
Gitoxin	gitoxigenin	$3\beta,14\beta,16\beta$-trihydroxy-5β-card-20(22)-enolide
Digoxin	digoxigenin	$3\beta,12\beta,14\beta$-trihydroxy-5β-card-20(22)-enolide
Periplocin	periplogenin	$3\beta,5\beta,14\beta$-trihydroxy-5β-card-20(22)-enolide
Sarmentocymarin	sarmentogenin	$3\beta,11\alpha,14\beta$-trihydroxy-5β-card-20(22)-enolide
k-Strophanthin	strophanthidin	$3\beta,5\beta,14\beta$-trihydroxy-19-oxo-5β-card-20(22)-enolide
Ouabain	ouabagenin	$1\beta,3\beta,5\beta,11\alpha,14\beta$,19-hexahydroxycard-20(22)-enolide

The most important cardiac glycosides, medicinally, are those occurring in foxglove (*Digitalis*): digitoxin, gitoxin, and digoxin. Each of these contains a specific aglycone (*e.g.*, digitoxigenin is the aglycone of digitoxin) linked to three molecules of the sugar digitoxose and is derived from a more complex glycoside (digilanides A, B, and C, respectively) from which glucose and acetic acid are removed during the isolation procedures.

Toxic glycosides

The squill, or sea onion, *Scilla maritima*, a seashore plant, contains several toxic glycosides of which the aglycones are bufadienolides more typical of the toad poisons than of plant products. (In a bufadienolide, two double bonds are present in the bufanolide side chain.) A white form of the squill produces the glycoside scillaren A, which contains the aglycone scillarenin, while a red form

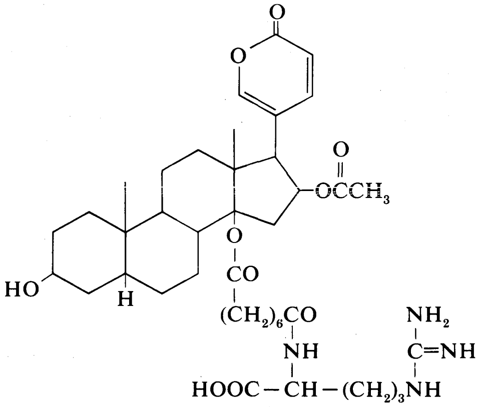


digitoxigenin (23)

produces scilliroside, which is specifically toxic to rodents and has long been used as a rat poison. The contribution of the side chain to cardiac activity differs little between the bufanolides and the cardanolides.

Toad poisons. Defensive venoms secreted by skin glands (principally the parotid glands) of the toad owe their high toxicity to bufadienolides that occur both free (bufogenins) and combined (bufotoxins). These compounds have digitalis-like properties and have been used medicinally in a traditional Chinese preparation, Ch'an Su. The best characterized is bufotoxin, from the European toad *Bufo vulgaris* and the Oriental toad *Bufo gargarizans,* of which the bufogenin is bufotalin, a close structural relative of gitoxin.

Sapogenins and saponins. Sapogenins are steroids of the spirostan type that occur widely and in great variety in plants. They are linked to sugars as glycosides, usually through a 3β-hydroxyl group. The glycosides are saponins, so called because they form soapy solutions and have other surface active (*e.g.,* hemolytic) properties. Since saponins are difficult to purify, the complete structures of only a few are known. Among these is dioscin (from certain yams, *Dioscorea* species, and yuccas, *Trillium* species); the steroid portion of this saponin is diosgenin.



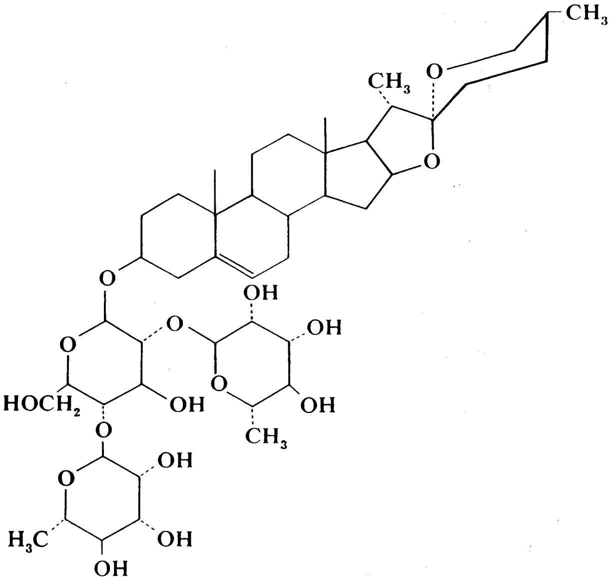
bufotoxin (24)

The nature and number of sugar residues per molecule are known for many saponins. These include the common sugars glucose, xylose, galactose, rhamnose, and arabinose. In most cases, however, only the structure of the sapogenin, which can be released from the saponin by acid hydrolysis, is known with certainty. Linkage of rings *A* and *B* may be *cis-* (5β) or *trans-* (5α) or may involve unsaturation at C-5. A hydroxyl group is nearly always present at position 3, and hydroxyl or ketonic groups may be present at positions 1, 2, 4, 5, 6, 11, 12, or 15. Many pairs of natural sapogenins differ only in configuration at C-25. Their structural features and abundance make diosgenin and hecogenin useful as starting materials for steroid hormone manufacture.

PHARMACOLOGICAL ACTIONS OF STEROIDS

ide
ffects

Aside from their principal physiological effects, all steroid hormones have generalized influences upon metabolic systems throughout the body. These are sometimes seen as powerful pharmacological side effects when, either during hormone therapy or through some endocrine abnormality, the body is exposed to excessive amounts of a naturally occurring steroid hormone. In some synthetic analogues of the natural hormones, a desired activity is accentuated while others are minimized. Furthermore, just as naturally occurring steroid hormones of differing biological activity (estrogens, androgens, glucocorticoids, mineralocorticoids) often act antagonistically, the many steroid analogues include a number of inhibitors of the natural hormones. Details of the effects of steroids upon living organisms, including pharmacological applications, are more fully covered in other articles (see BIOCHEMICAL COMPONENTS OF ORGANISMS; ENDOCRINE SYSTEMS).



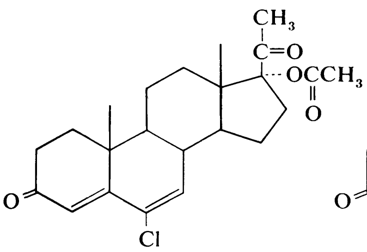
dioscin (25), a saponin of diosgenin

Androgens and anabolic steroids. Androgens secreted or administered in abnormally large amounts can cause development of male characteristics in the female and precocious sexual development in the male. Conversely, hypogonadism of the male (inadequate testicular function) leads to retarded sexual development and retention of feminine bodily characteristics (eunuchoidism), which can sometimes be remedied by administration of androgenic steroids. Several esters of testosterone are commonly used by injection for this purpose. Many orally active analogues of testosterone are also available (Table 51) in which activity is greatly enhanced, and often the ratio of androgenic activity to anabolic activity is shifted markedly in favour of the latter. This ratio primarily determines the therapeutic value of these compounds as anabolic agents. They are used together with growth hormone to promote growth in children in whom growth is retarded. They are also used to promote physical recovery from debilitating diseases. Their alleged use by some athletic competitors in sports has been decried.

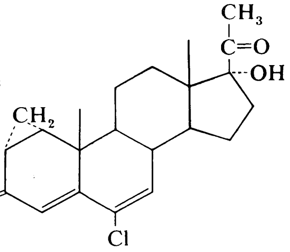
Use as
anabolic
agents

The psychic effects of androgens have been exploited in their use to correct frigidity in women and lack of self-esteem in adolescents. Androgens are also used in the treatment of estrogen-dependent tumours, such as breast cancer in women.

Anti-androgens. The estrogens and a number of the syn-



chlormadinone acetate (26)



cyproterone (27)

Table 51: Relative Oral Androgenic and Anabolic Activities of Testosterone and Some Synthetic Analogs

nonproprietary name (trade name)	systematic name	relative potency	
		androgenic*	anabolic†
Testosterone	17β-hydroxy-4-androsten-3-one	100	100
Methyltestosterone (Metandren, Oreton-M)	17β-hydroxy-17-methyl-4-androsten-3-one	200	200
Methandrostenolone (Dianabol)	17β-hydroxy-17-methyl-1, 4-androstadien-3-one	120	400
Norethandrolone (Nilevar)	17α-ethyl-17-hydroxy-4-estren-3-one	40	400
Stanozolol (Winstrol)	17β-hydroxy-17-methyl-5α-androstano(3, 2-*c*)pyrazole	60	600
Oxymetholone (Adroyd, Anadrol)	2-hydroxymethylene-17β-hydroxy-17-methyl-5α-androstan-3-one	100	1,000
Fluoxymesterone (Halotestin, Ultandren)	9α-fluoro-11β, 17-dihydroxy-17α-methyl-4-androsten-3-one	1,500	3,500

*Seminal vesicle weight response. †*Levator ani* weight response.

thetic gestogens, such as medroxyprogesterone acetate and chlormadinone acetate (Table 52), have anti-androgenic properties that are the basis for their use against benign or malignant hyperplasia of androgen-dependent tissues such as the prostate. Other anti-androgens are A-nortestosterone and A-norprogesterone and their derivatives. The anti-androgens cyproterone and medroxyprogesterone acetate have been used successfully to control sexually violent patients.

The Pill

Synthetic estrogens and gestogens: steroid contraceptives. The most effective currently available method of contraception uses combinations of synthetic estrogens and gestogens that prevent ovulation and render the uterine environment inimical to conception and to the development of the ovum. Unlike the natural hormones, these synthetic steroids are highly active when taken orally, as tablets, and as contraceptives they are popularly referred to as the Pill. Ovulation is suppressed because release of the pituitary hormone that stimulates it is inhibited, primarily by the estrogenic component. The gestogens are also anti-ovulatory, but their principal function in oral contraceptives is to control the uterine bleeding (menstruation) that follows withdrawal of the estrogen. The tablets are taken according to schedules that simulate a 28-day menstrual cycle. Most frequently, a combination of the two types of steroid is taken daily during 21 or 22 days and is then interrupted for the remainder of the 28-day period. During the nonmedication days, the menstrual flow occurs. An alternative sequential schedule aims to approximate more closely to the endocrine events of the normal menstrual cycle by the use of tablets of estrogen alone for 14–15 days followed by tablets of combined estrogen and gestogen for 5–7 days prior to interruption to permit menstruation. The combination method is somewhat more reliable, probably because the gestogens in the combination regime influence the uterine environment to prevent implantation of the ovum in those few cycles in which ovulation is not inhibited.

The contraceptive steroids, which were brought into popular use in the early 1960s, have proved remarkably safe. Their discontinuance is usually followed by a prompt return of fertility. Some individuals suffer unpleasant symptoms (nausea, water retention, tiredness) in the use of certain preparations, but a variety of alternative types have become available (Table 52), among which a satisfactory one usually may be found. Estrogens that have been used in oral contraceptives include estranol, which has about the same potency as estradiol, and mestranol, which is less potent. The main differences between preparations are the character of the gestogens and the quantities and ratios of the steroids that they contain.

Other uses of estrogens and gestogens

Besides their use in birth control, the contraceptive steroid preparations and synthetic gestogens are used to correct irregularities of the menstrual cycle and its frequent physical and psychological symptoms. The gestogens are used to maintain pregnancy in cases of threatened abortion, some as anti-androgens, and others as anti-estrogens. Contraceptive steroid preparations are also used for control of breeding cycles and synchronization of estrus in farm animals. The synthetic estrogens are used to alleviate the unpleasant physical and psychic symptoms of the menopause in women and for the treatment of some cancers of the prostate in men and of the breast in women. They are also used in some cases to suppress the plasma cholesterol concentrations in sufferers from advanced atherosclerosis, a disease that is markedly more prevalent in men and postmenopausal women than in younger women with intact ovaries.

Anti-estrogens. Androgenic steroids are anti-estrogenic and are often used to suppress the growth of estrogen-dependent cancers. Methyltestosterone and fluoxymesterone (Table 51), and norethindrone (norethisterone) and norgestrel (Table 52) are anti-estrogens. Since estrogens block the release of the pituitary hormone responsible for ovulation, a potent anti-estrogen can stimulate ovulation by inhibiting this action of estrogens.

Adrenal hormones. The anti-inflammatory and glucocorticoid activities of cortisol are enhanced, in some cases with relative reduction of its mineralocorticoid activity, by various structural modifications (Table 53). A 9α-fluorine

Table 52: Oral Gestogens and Contraceptive Preparations

nonproprietary name (trade name)	systematic name	proprietary name of contraceptive*
17α-Hydroxyprogesterone caproate (Delalutin)	17α-hydroxy-4-pregnene-3, 20-dione caproate	
Medroxyprogesterone acetate (Provera)	17α-hydroxy-6α-methyl-4-pregnene-3, 20-dione acetate	Provest (E)
Megestrol acetate	17α-hydroxy-6-methyl-4, 6-pregnadiene-3, 20-dione acetate	Volidan (E)
Chlormadinone acetate	6-chloro-17α-hydroxy-4, 6-pregnadiene-3, 20-dione acetate	C-Quens (M)†
Dimethisterone	6α, 21-dimethyl-17α-ethynyl-4-androsten-3-one	Oracon (E)† Ovin (E)
Norethindrone, norethisterone (Norlutin)	17α-ethynyl-17-hydroxy-4-estren-3-one	Ortho-Novum (M)‡ Norinyl (M) Norquen (M)†
Norethynodrel	17α-ethynyl-17-hydroxy-5(10)-estren-3-one	Enovid, Enovid-E (M)
Lynestrenol	17α-ethynyl-4-estren-17-ol	Lyndiol (M)
Norethindrone acetate	17α-ethynyl-17-hydroxy-4-estren-3-one acetate	Anovlar 21, Norlestrin (E)
Norgestrel	18-methyl-17α-ethynyl-17-hydroxy-4-estren-3-one	Ovral (E)
	17α-ethynyl-17-hydroxy-4, 9(10), 11(12)-estratrien-3-one	Planor (M)
	3-cyclopentoxy-17α-ethynyl-3, 5-estradien-17-ol acetate	Riglovis (E)
Ethynodiol diacetate	17α-ethynyl-4-estrene-3β, 17β-diol diacetate	Ovulen (M)

*The estrogen used is denoted by (E), estranol or (M), mestranol. †Sequential. ‡Sequential or combination.

Table 53: Approximate Relative Anti-inflammatory Potency of Some Analogs of Cortisol

nonproprietary name	modification of cortisol structure	clinically estimated dose (mg) equivalent to 20 mg. cortisol
Cortisone	11-dehydro-	25
Prednisone	1, 11-dehydro-	5
Prednisolone	1-dehydro-	5
Methylprednisolone	1-dehydro-, 6α-methyl-	4
Triamcinolone	1-dehydro-, 9α-fluoro-, 16α-hydroxy-	4
Fluprednisolone	1-dehydro-, 6α-fluoro-	2
Paramethasone	1-dehydro-, 6α-fluoro-, 16α methyl-	2
Triamcinolone acetonide	1-dehydro-, 9α-fluoro-, 16α-hydroxy-, 16, 17-acetonide	1
Dexamethasone	1-dehydro-, 9α-fluoro-, 16α-methyl-	0.75
Betamethasone	1-dehydro-, 9α-fluoro-, 16β-methyl-	0.6
Fluocinolone acetonide	1-dehydro-, 6α, 9α-difluoro-, 16α-hydroxy-, 16, 17-acetonide	0.5
Fluandrenolone acetonide	6α-fluoro-, 16α-hydroxy-, 16, 17-acetonide	0.4

atom enhances the glucocorticoid activity of cortisol about 10-fold, but its salt-retaining activity about 50-fold. On the other hand, unsaturation at C-1 increases glucocorticoid, but not mineralocorticoid, activity, and 6α-fluorine or methyl and 16-methyl or hydroxyl groups (and especially 16α,17α-acetonides—*i.e.,* compounds formed from 16α,17α-dihydroxy compounds and acetone) enhance anti-inflammatory activity while reducing salt activity. These groupings, therefore, appear in various combinations in anti-inflammatory steroids, many of which, however, lack the salt-retaining activity necessary for total adrenal replacement therapy. Cortisol analogues, such as dexamethasone, are used to treat many inflammatory and rheumatic diseases, to suppress the immune response in allergies and in organ transplantation, and to delay the progress of leukemia. They are also widely used for the treatment of local inflammatory reactions. A synthetic steroid of a quite different type, spironolactone (Aldactone A), is used as an antagonist to the action of aldosterone in certain cases of hypertension.

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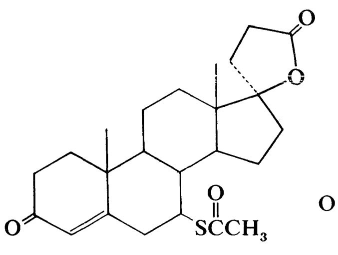
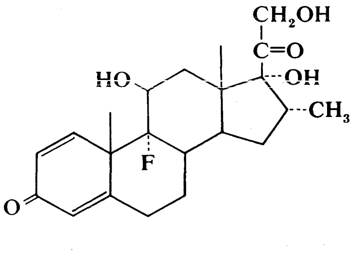


spironolactone (28)



dexamethasone (29)

Cardiotonic steroids. Preparations in which cardiotonic steroids of both vegetable and animal origin are the active principles have been used as emetics, diuretics, and arrow poisons since ancient times. The use of digitalis, ouabain, and strophanthin glycosides to slow the rate and strengthen the contractility of the failing heart is one of the most important methods of treatment of this condition. Of these agents, the digitalis glycosides are the most widely used. The therapeutic effects of these agents are related to their influence upon muscle cells, but knowledge of their mechanism of action is incomplete and their use in medicine remains essentially empirical.

Conclusion: The general biological significance of steroids. That such diverse physiological functions and effects should be shown by steroids, all of which are synthesized by essentially the same central biosynthetic pathway, is a remarkable example of biological economy. Most of these functions, especially those of a hormonal type, involve the transmission of biologically essential information. The specific information content of the steroid resides in the character and arrangement of its substituent groups and in other subtle structural modifications. The mechanism of action of no steroid hormone at the molecular level is as yet known, nor is it known why the steroid nucleus is so highly appropriate as a carrier of informational signals in systems of biological control. But that it is, is beyond doubt, and it seems reasonable to expect that many further regulatory functions of steroids, especially in lower life forms, remain to be discovered.

(R.B.C./Ed.)

COORDINATION COMPOUNDS

Coordination compounds are substances with characteristic chemical structures in which a central metal atom is surrounded by nonmetallic atoms (or groups of atoms), called ligands, joined to it by chemical bonds. The class includes a number of important biological materials, such as vitamin B$_{12}$ and hemoglobin, the red colouring matter of blood. It also includes a number of industrially important materials used as dyestuffs and pigments; as agents for extracting, purifying, and analyzing metals; and as catalysts for preparing such useful organic substances as the polyethylene plastics. Coordination compounds have been much studied because of what they reveal about molecular structure and chemical bonding, as well as because of the unusual chemical nature and the useful properties of certain coordination compounds.

The general class of coordination compounds—or complexes, as they are sometimes called—is extensive and diverse. The substances in the class may be composed of electrically neutral molecules or of positively or negatively charged species (ions).

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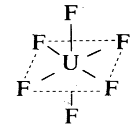
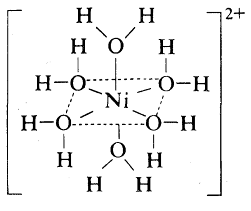
ion com-
ounds

Among the many coordination compounds having neutral molecules is uranium hexafluoride. The molecular formula of this compound, UF$_6$, indicates that six atoms of fluorine (F) are joined to one atom of uranium (U). The structural formula of the compound represents the actual arrangement of atoms in the molecules:



In this formula the solid lines, which represent bonds between atoms, show that four of the fluorine atoms are bonded to the single atom of uranium and lie in a plane with it, the plane being indicated by dotted lines (which do not represent bonds); whereas the remaining two fluorine atoms (also bonded to the uranium atom) lie above and below the plane, respectively.

An example of an ionic coordination complex is the hydrated (water-containing, with formula H$_2$O) ion of nickel, symbol Ni, molecular formula [Ni(H$_2$O)$_6$]$^{2+}$, the structure of which is

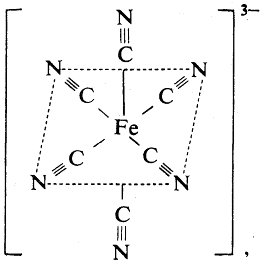


in which the symbols and lines are used as above and the brackets and the "two plus" (2+) sign show that the double positive charge goes with the whole unit.

The central metal atom in a coordination compound itself may be neutral or charged (ionic). The coordinated groups—or ligands, as they are usually called—may be neutral molecules such as water (in the above example), ammonia, or carbon monoxide; negatively charged ions (anions) such as fluoride (in the first example above) or cyanide ion; or, occasionally, positively charged ions (cations) such as hydrazinium or nitrosonium ion.

Complex ions

Complex ions—that is, the ionic members of the family of coordination substances—may exist as free ions in solution, or they may be incorporated into crystalline materials (salts) with other ions of opposite charge. In such salts, the complex ion may be either the cationic (positively charged) or the anionic (negatively charged) component (or, on occasion, both). The hydrated nickel ion (above) is an example of a cationic complex. An anionic complex is the hexacyanide of ferric (symbol Fe) ion, $[Fe(CN)_6]^{3-}$, or



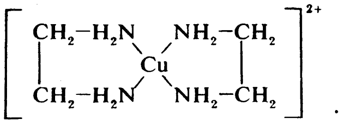
in which the symbols have the same general meaning as in the above examples. Crystalline salts containing these ions are potassium ferricyanide $K_3[Fe(CN)_6]$, and the hexahydrate of nickel chloride, $[Ni(H_2O)_6]Cl_2$, in each case the charge on the complex ion being neutralized by appropriate numbers of ions of opposite charge. In the case of potassium ferricyanide, three positively charged potassium ions, K^+, neutralize the charge on the complex, and in the nickel complex the charges are neutralized by negative chloride ions, Cl^-.

Chelates and chelation

Each molecule or ion of a coordination compound includes a number of ligands, and, in any given substance, the ligands may be all alike or they may be different. Attachment of the ligands to the metal atom may be through only one bond or it may be through several bonds. When only one bond is involved, the ligand is said to be unidentate; when two are involved, it is bidentate; and so on. In general, ligands utilizing more than one bond are said to be multidentate. Because a multidentate ligand is joined to the metal atom in more than one place, the resulting complex is said to be cyclic—*i.e.,* to contain a ring of atoms. Coordination compounds containing multidentate ligands are called chelates (Greek *chele,* "claw"), and their formation is termed chelation. Chelates are particularly stable and useful (see below *Multidentate*). An example of a typical chelate is the complex formed between cupric ion (Cu^{2+}) and the organic compound ethylenediamine (formula $NH_2CH_2CH_2NH_2$). The formula of the complex is

$$[Cu(NH_2CH_2CH_2NH_2)_2]^{2+},$$

the structural formula is



The total number of bonds from the ligands to the metal atom is called the coordination number; generally, coordination numbers range between 2 and 12, with 4 and 6 being the most common. The ligands surrounding the central atom often are referred to collectively as the coordination sphere.

The distinction between coordination compounds and other substances is, in fact, somewhat arbitrary. The designation coordination compound, however, is generally restricted to substances the molecules or ions of which are discrete entities and in which the central atom is metallic. Accordingly, molecules such as sulfur hexafluoride and carbon tetrafluoride are not normally considered to be coordination compounds because sulfur and carbon are nonmetallic elements. Yet there is no great difference between these compounds and, say, uranium hexafluoride. Furthermore, such simple ionic salts as sodium chloride or nickel difluoride are not considered to be coordination compounds because they consist of continuous ionic lattices rather than discrete molecules. Nevertheless, the arrangement (and bonding) of the anions surrounding the metal ions in these salts is similar to that in coordination compounds. Coordination compounds generally are characterized by a variety of distinctive physical and chemical properties, such as colour, magnetic susceptibility, solubility and volatility, an ability to undergo oxidation-reduction reactions, and catalytic activity (see below).

GENERAL ASPECTS

History. Among the earliest recorded examples of a coordination compound is the substance Prussian blue, with formula $Fe_4[Fe(CN)_6]_3$, which has been used as an artist's pigment since the beginning of the 18th century. Another early example of the preparation of a coordination compound is the use of a sparingly soluble compound, potassium hexachloroplatinate, formula $K_2[PtCl_6]$, in 1760 to refine the element platinum.

The sustained and systematic development of modern coordination chemistry, however, usually is considered to have begun with the discovery in 1798 that ammoniacal solutions of a cobalt chloride, $CoCl_3$, on standing overnight, deposited orange crystals with the composition $CoCl_3 \cdot 6NH_3$, the correct formulation of which is now recognized to be $[Co(NH_3)_6]Cl_3$; this shows that the six ammonia groups are associated with the cobalt atom and the whole is neutralized by three chlorine atoms. The feature of this observation that was particularly significant was the recognition that two independently stable compounds (*i.e.,* cobalt chloride and ammonia) could combine to form a new chemical compound with properties quite different from those of the constituent compounds.

Formulations of Werner

Discoveries of further cobaltammine complexes and many other coordination compounds were made in the 19th century, but it was not until 1893 that Alfred Werner, a Swiss chemist, proposed the first essentially correct formulation of such compounds and provided the first insight into the nature of their bonding. Werner reformulated a series of platinum complexes and showed that their dissociation into ions in solution took place as shown in Table 54. In each case, the number of ions formed was determined by measurements of the electrical conductivity of the solutions and by analytical determinations of the free chloride ions (where these were formed). It is noteworthy that in Werner's formulations the coordination number of platinum retains the constant value of 6. This constancy of coordination number of metal atoms occurs frequently, though not invariably.

Table 54: Platinum Complexes as Formulated by Werner

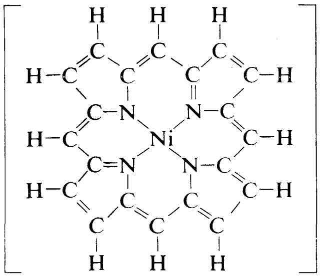
old formulation	Werner formulation	dissociation
$PtCl_4 \cdot 6NH_3$	$[Pt(NH_3)_6]Cl_4$	$\rightarrow [Pt(NH_3)_6]^{4+} + 4Cl^-$
$PtCl_4 \cdot 5NH_3$	$[Pt(NH_3)_5Cl]Cl_3$	$\rightarrow [Pt(NH_3)_5Cl]^{3+} + 3Cl^-$
$PtCl_4 \cdot 4NH_3$	$[Pt(NH_3)_4Cl_2]Cl_2$	$\rightarrow [Pt(NH_3)_4Cl_2]^{2+} + 2Cl^-$
$PtCl_4 \cdot 3NH_3$	$[Pt(NH_3)_3Cl_3]Cl$	$\rightarrow [Pt(NH_3)_3Cl_3]^+ + Cl^-$
$PtCl_4 \cdot 2NH_3$	$[Pt(NH_3)_2Cl_4]$	→ none
$PtCl_4 \cdot NH_3 \cdot KCl$	$K[Pt(NH_3)Cl_5]$	$\rightarrow K^+ + [Pt(NH_3)Cl_5]^-$
$PtCl_4 \cdot 2KCl$	$K_2[PtCl_6]$	$\rightarrow 2K^+ + [PtCl_6]^{2-}$

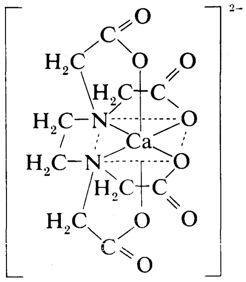
Classification. A coordination compound is characterized by the nature of the central metal atom or ion; the oxidation state of the latter (that is, the gain or loss of electrons in passing from the neutral atom to the charged ion); and by the number, kind, and arrangement of the ligands. Because virtually all metallic elements form coordination compounds, sometimes in several oxidation states and usually with many different ligands, a large number of coordination compounds is known.

Mononuclear, unidentate. The simplest types of coordination compounds are those containing a single metal atom or ion (mononuclear compounds), surrounded by unidentate ligands. Most of the coordination compounds already cited, such as the platinum complexes of Werner, belong to this class. Among the ligands forming such complexes are a wide variety of neutral molecules, such as ammonia, water, carbon monoxide, and nitrogen, as well as monoatomic and polyatomic anions, such as the hydride, fluoride, chloride, oxide, hydroxide, nitrite, thiocyanate, carbonate, sulfate, and phosphate ions. Coordination of such ligands to the metal virtually always occurs through an atom possessing an unshared pair of electrons, which it donates to the metal to form a coordinate bond with the latter. Among the atoms that are known to coordinate to metals are those of virtually all the nonmetallic elements (such as hydrogen, carbon, oxygen, nitrogen, and sulfur), with the exception of the noble gases (helium, neon, argon, krypton, and xenon).

Coordination via unshared electron pair

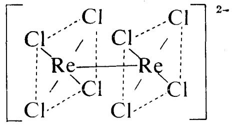
Multidentate. The chelate complex of a copper ion and ethylenediamine mentioned above is an example of a compound formed between a metal ion and a bidentate ligand. Two further examples of chelate complexes are shown below. These are a nickel complex with a tetradentate large-ring ligand, known as a porphyrin, and a calcium complex with a hexadentate ligand, ethylenediaminetetraacetate. Because metal-ligand attachment in such chelate complexes is through several bonds, such complexes tend to be very stable.


nickel-porphyrin complex

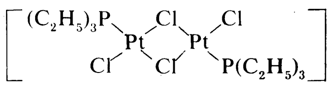


calcium-ethylenediamine-tetraacetate complex

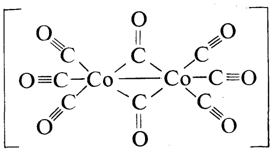
Certain ligands may be either unidentate or multidentate, depending on the particular compound in which they occur. The carbonate ion, formula $(CO_3)^{2-}$, for example, is coordinated to the cobalt ions in two cobalt compounds, $[Co(NH_3)_5(CO_3)]^+$ and $[Co(NH_3)_4(CO_3)]^+$, through one and two oxygen atoms, respectively.

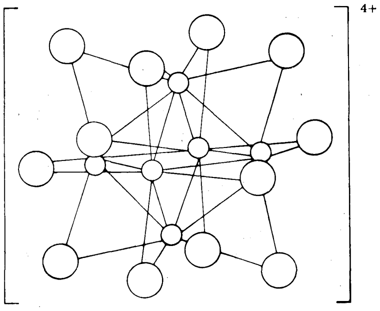


complex with metal-metal bond (Re is rhenium)



complex with bridging ligands (C_2H_5 is ethyl radical, Pt is platinum, P is phosphorus, and Cl is chlorine)


complex with metal-metal bond and bridging ligands



complex with metal-cluster nucleus

○ = molybdenum atoms

◯ = chlorine atoms

Polynuclear. Polynuclear complexes are coordination compounds containing two or more metal atoms, or ions, in a single coordination sphere. The two atoms may be held together through direct metal–metal bonds, through bridging ligands, or both. Examples of each are shown below, along with a unique metal-cluster complex having six metal atoms in its nucleus.

Nomenclature. Generally, the systematic naming of coordination compounds is carried out by rules recommended by the International Union of Pure and Applied Chemistry (IUPAC). Among the more important of these are the following:

IUPAC rules of nomenclature

1. Neutral and cationic complexes are named by first identifying the ligands, followed by the metal and its oxidation number (in Roman numerals enclosed within parentheses). Anionic ligands are cited ahead of neutral ones, and inorganic ligands ahead of organic ones. When the complex contains more than one ligand of a given kind, the number of such ligands is designated by one of the prefixes di-, tri-, tetra-, penta-, and so on, or, in the case of complex ligands, bis-, tris-, tetrakis-, pentakis-, and so on. The oxidation number of the metal is defined in the customary way as the residual charge on the metal if all the ligands were removed together with the electron pairs involved in coordination to the metal. The following examples are illustrative:

$[Co(NH_3)_5Cl]^{2+}$	chloropentaamminecobalt(III)
$[Ni(NH_2CH_2CH_2NH_2)_3]^{2+}$	tris(ethylenediamine)nickel(II)
$[Fe(CO)_5]$	pentacarbonyliron(0)
$[Al(H_2O)_5(OH)]^{2+}$	hydroxopentaaquo-aluminum(III)
$[RhH(CO)(P[C_6H_5]_3)_3]$	hydridocarbonyltris(triphenyl-phosphine)rhodium(I).

2. Anionic complexes are similarly named, except that the name is terminated by the suffix -ate; *e.g.,*

$[PtCl_4]^{2-}$	tetrachloroplatinate(II)
$[Ag(S_2O_3)_2]^{3-}$	bis(thiosulfato)argentate(I)
$[OsNCl_5]^{2-}$	nitridopentachloroosmate(VI)
$[Cr(SCN)_4(NH_3)_2]^-$	tetrathiocyanatodiammine-chromate(III).

3. In the case of salts, the cation is named first, then the anion; *e.g.,*

$[Pt(NH_3)_3Cl]Cl$	chlorotriammineplatinum(II) chloride
$K_3[Fe(CN)_6]$	potassium hexacyanoferrate(III)
$[Cr(NH_3)_6][Co(CN)_6]$	hexaamminechromium(III) hexacyanocobaltate(III).

4. Polynuclear complexes are named as follows, bridging ligands being identified by a prefix consisting of the Greek letter mu (μ-):

$[(CO)_5Mn-Mn(CO)_5]$	decacarbonyldimanganese (0) *or* bis(pentacarbonyl-manganese)
$[(NH_3)_5Cr-OH-Cr(NH_3)_5]Cl_5$	μ-hydroxo-bis[pentaamminechromium(III)] chloride

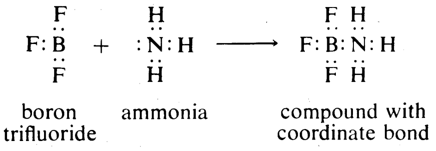
In addition to their systematic designations, many coordination compounds are also known by names reflecting their discoverers or colours. Examples are:

$K[PtCl_3(C_2H_4)]$	Zeise's salt
$NH_4[Cr(NH_3)_2(SCN)_4] \cdot H_2O$	Reinecke's salt
$[Co(NH_3)_5(H_2O)]Cl_3$	roseocobaltic chloride (red)
$Fe_4[Fe(CN)_6]_3$	Prussian blue

Coordinate bonding. Werner originally postulated that coordination compounds could be formed because the central atoms carried the capacity to form secondary or

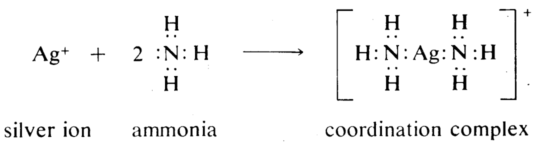
Electron pair theory of bonding

coordinate bonds, in addition to the normal or valence bonds. A more complete description of coordinate bonding, in terms of electron pairs, became possible in the 1920s following the introduction of the concept that all covalent (nonionic) bonds consist of electron pairs shared between atoms, a concept advanced chiefly by the U.S. chemist Gilbert N. Lewis. In Lewis' formulation, when both electrons are contributed by one of the atoms, as in the boron–nitrogen bond, which is formed when the substance boron trifluoride combines with ammonia, the bond is called a coordinate bond:



In these formulas, the valence (or bonding) electrons are indicated by dots, each pair of dots between two atomic symbols representing a bond between the corresponding atoms.

Following Lewis' ideas, the suggestion was made that the bonds between metals and ligands were of this same type, with the ligands acting as electron donors and the metal ions as electron acceptors. This suggestion provided the first electronic interpretation of bonding in coordination compounds. The coordination reaction between silver ions and ammonia illustrates the resemblance of coordination compounds to the situation in the boron–nitrogen compound. According to this view, the metal ion can be regarded as a so-called Lewis acid and the ligands as Lewis bases:



In the above formulas, the dots again represent valence electrons.

Considerable success in understanding certain coordination compounds also has been achieved by treating them as examples of simple ionic or electrostatic bonding. This view attributes the bonding in coordination compounds to electrostatic forces between the positively charged metal ions and negatively charged ligands or, alternatively, in the case of neutral ligands (*e.g.,* water and ammonia), charge separations (dipoles) that appear within the molecules. Although this approach meets with considerable success for complexes of metal ions with small electronegative ligands, such as fluoride or chloride ions or water molecules, it breaks down for ligands of low polarity (charge separation), such as carbon monoxide. It also requires modification to explain why the spectral (light absorption) and magnetic properties of coordinated metal ions generally differ from those of the free ions and, for a given metal ion, depend on the nature of the ligands. Thus, the complex ion hexafluoroferrate(III) (FeF_6^{3-}), has magnetic properties to be expected from a substance with five unpaired electrons, as does the free iron(III) ion (Fe^{3+}), whereas the magnetic properties of the closely related hexacyanoferrate(III) $[Fe(CN)_6]^{3-}$ correspond to only one unpaired electron.

The ligand-field theory

Since 1950 it has been apparent that a more complete theory, which incorporates contributions from both ionic and covalent bonding, is necessary to give an adequate account of the properties of coordination compounds. Such a theory is the so-called ligand-field theory, which has its origin in the more general theory of chemical bonding called the molecular-orbital theory (molecular orbitals describing the spatial distributions of electrons in molecules, just as atomic orbitals describe the distributions in atoms). This theory accounts with remarkable success for most properties of coordination compounds.

An important conclusion from ligand field theory is that two types of bonds, called sigma bonds and pi bonds, occur in coordination compounds just as they do in ordinary covalent (organic) compounds. Sigma bonds are the more usual of the two, and they are symmetrical about the axis of the bond; pi bonds, which are less common, are unsymmetrical with regard to the bond axis. In coordination compounds, pi bonding may result from donation of electrons from ligands, such as fluorine or oxygen atoms, to empty *d* orbitals of the metal atoms (the designation *d* merely being a way of indicating the particular orbitals involved). An example of this type of bonding is the chromate ion, $(CrO_4)^{2-}$, in which the oxygen atoms donate electrons to the central chromium ion. Alternatively, electrons from *d* orbitals of the metal atom may be donated to empty orbitals of the ligand. This is the case in the compound tetracarbonylnickel, in which empty pi orbitals in the carbon monoxide molecules accept *d*-orbital electrons from the nickel atom.

Coordination number and geometry. Among the essential properties of coordination compounds are the number and arrangement of the ligands attached to the central metal atom or ion—that is, the coordination number and the coordination geometry, respectively. The coordination number of a particular complex is determined by the relative sizes of the metal atom and the ligands, by spatial (steric) constraints governing the shapes (conformations) of multidentate ligands, and by electronic factors, most notably the electronic configuration of the metal ion. The influence of the latter is illustrated by the examples in Table 55. The numbers labelled "total number of valence electrons" in this table comprise the *d* electrons of the metal ion together with the pair of electrons donated by each of the ligands.

Table 55: Coordination Numbers and Geometries of Metal Cyanide Complexes

electron configuration*	metal ion	cyanide complex	geometry	total number of valence electrons
d^2	Mo^{4+}	$[Mo(CN)_8]^{4-}$	dodecahedral	18
d^3	Cr^{3+}	$[Cr(CN)_6]^{3-}$	octahedral	15
d^4	Mn^{3+}	$[Mn(CN)_6]^{3-}$	octahedral	16
d^5	Fe^{3+}	$[Fe(CN)_6]^{3-}$	octahedral	17
d^6	Co^{3+}	$[Co(CN)_6]^{3-}$	octahedral	18
d^7	Co^{2+}	$[Co(CN)_5]^{3-}$	square pyramidal	17
d^8	Ni^{2+}	$[Ni(CN)_4]^{2-}$	square planar	16
d^8	Ni^{2+}	$[Ni(CN)_5]^{3-}$	square pyramidal or trigonal bipyramidal	18
d^{10}	Cd^{2+}	$[Cd(CN)_4]^{2-}$	tetrahedral	18
d^{10}	Ag^+	$[Ag(CN)_2]^-$	linear	14

*Number of *d* electrons indicated by superscript.

The 18-electron rule

Contributing to the pattern of coordination numbers in Table 55 is the 18-electron rule (sometimes called the noble gas rule), which states that coordination compounds in which the total number of valence electrons approaches but does not exceed 18 (the number of electrons in the valence shells of the noble gases) are most stable. The stabilities of 18-electron valence shell's are also reflected in the coordination numbers of the stable mononuclear carbonyls of different metals—*e.g.,* tetracarbonylnickel, pentacarbonyliron, and hexacarbonylchromium (each of which has a valence shell of 18).

The 18-electron rule applies particularly to covalent complexes, such as the cyanides, carbonyls, and phosphines. For more ionic (also called outer-orbital) complexes, such as fluoro or aquo complexes, electronic factors are less important in determining coordination numbers, and configurations corresponding to more than 18 valence shell electrons are not uncommon. Several nickel complexes, for example, including the hexafluoro, hexaaquo, and hexaammine complexes, each have 20 valence shell electrons.

Any one metal ion tends to have the same coordination number in different complexes—*e.g.,* generally 6 for chromium(III)—but this is not invariably so. Differences in coordination number, for example, may result from difference in the sizes of the ligands. The iron(III) ion, for example, is able to accommodate six fluoride ions in the hexafluoride complex, but only four of the larger chloride ions in the tetrachloride. In some cases, a metal ion and ligand form two or more complexes with different coordination numbers—*e.g.*, tetracyanonickelate(II) and pentacyanonickelate(II).

Isomerism. Coordination compounds often exist as isomers—*i.e.*, as compounds with the same chemical composition but different structural formulas. Many different kinds of isomerism occur among coordination compounds. The following are some of the more common types.

Ionization isomerism. Certain isomeric pairs occur that differ only in that two ionic groups exchange positions within (and without) the primary coordination sphere. These are called ionization isomers and are exemplified by the two compounds $[Co(NH_3)_5Br]SO_4$ and $[Co(NH_3)_5(SO_4)]Br$, in the first of which the bromide ion is coordinated to the cobalt ion, and the sulfate ion is outside the coordination sphere; and in the second of which the sulfate ion occurs within the coordination sphere, and the bromide ion is outside it.

Coordination isomerism. Ionic coordination compounds in which both the positive and negative ions are ionic can exist as isomers if the ligands associated with the two metal atoms are exchanged, as in the pair of compounds shown by the formulas $[Co(NH_3)_6][Cr(CN)_6]$ and $[Cr(NH_3)_6][Co(CN)_6]$. Such compounds are called coordination isomers, as are the isomeric pairs obtained by redistributing the ligands between the two metal atoms, as in the doubly coordinated pair, $[Pt(NH_3)_4][PtCl_6]$ and $[Pt(NH_3)_4Cl_2][PtCl_4]$.

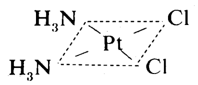
Ligand isomerism. Isomeric coordination compounds are known in which the overall isomerism results from isomerism solely within the ligand groups. An example of such isomerism is shown by the ions

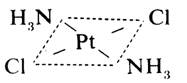
$[Pt(NH_2CH_2CH_2CH_2NH_2)_2]^{2+}$ and

$$[Pt(NH_2\underset{\displaystyle CH_3}{\underset{|}{C}}HCH_2NH_2)_2]^{2+}.$$

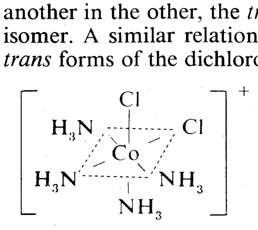
Linkage isomerism. Isomerism also results when a given ligand is joined to the central atom through different atoms of the ligand. Such isomerism is called linkage isomerism. A pair of linkage isomers are the ions $[Co(NH_3)_5(NO_2)]^{2+}$ and $[Co(NH_3)_5(ONO)]^{2+}$, in which the nitro group is joined to the cobalt atom through nitrogen and oxygen atoms, as shown by designating the nitro group by the formulas NO_2^- and ONO^-, respectively. Another example of this variety of isomerism is given by the pair of ions $[Co(CN)_5(NCS)]^{3-}$ and $[Co(CN)_5(SCN)]^{3-}$, in which an isothiocyanate $(NCS)^-$, and a thiocyanate group $(SCN)^-$, are bound to the cobalt ion through a nitrogen and a sulfur atom, respectively.

Geometrical isomerism. Geometrical isomers of coordination compounds differ from one another only in the manner in which the ligands are distributed; for example, in the isomeric pair of dichlorodiammineplatinum(II) compounds

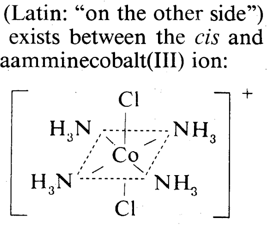




the two ammonia molecules and the two chlorine atoms are situated next to another in one isomer, called the *cis* (Latin: "on this side") isomer, and across from one another in the other, the *trans* (Latin: "on the other side") isomer. A similar relationship exists between the *cis* and *trans* forms of the dichlorotetraamminecobalt(III) ion:

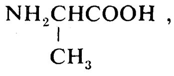


cis



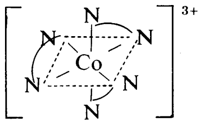
trans

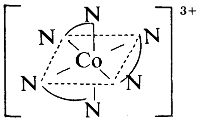
Optical isomerism. Optical isomers possess the ability to rotate plane-polarized light in opposite directions (a property determined with an instrument called a polarimeter). Optical isomers exist when the molecules of the substances are mirror images but are not superimposable upon one other. In coordination compounds, optical isomerism can arise either from the presence of an asymmetric ligand, such as one isomer of an amino acid,



$$NH_2\underset{\displaystyle CH_3}{\underset{|}{C}}HCOOH\ ,$$

or from an asymmetric arrangement of the ligands. Familiar examples of the latter variety are octahedral complexes carrying three bidentate ligands, such as ethylenediamine, $NH_2CH_2CH_2NH_2$. The two optical isomers corresponding to such a complex are depicted by the structures below:





For convenience, the ethylenediamine ligands above are indicated merely by a curved line between the symbols for the nitrogen (N) atoms.

PRINCIPAL CLASSES

The tendency for complexes to form between a metal ion and a particular combination of ligands and the properties of the resulting complexes depend on a variety of properties of both the metal ion and the ligands. Among the pertinent properties of the metal ion are its size, charge, and electron configuration. Relevant properties of the ligand include its size and charge, the number and kinds of atoms available for coordination, the sizes of the resulting chelate rings formed (if any), and a variety of other geometrical (steric) and electronic factors.

Stabilization of oxidation states

Many elements, notably certain of the metals, exhibit a range of oxidation states—that is, they are able to gain or lose varying numbers of electrons. The relative stabilities of these oxidation states are markedly affected by coordination of different ligands. The highest oxidation states correspond to empty or nearly empty *d* subshells (as the patterns of *d* orbitals are called). These states are generally stabilized most effectively by small negative ligands, such as fluorine and oxygen atoms, which possess unshared electron pairs. Such stabilization reflects, in part, the contribution of pi bonding caused by electron donation from the ligands to empty *d* orbitals of the metal ions in the complexes. Conversely, neutral ligands, such as carbon monoxide and unsaturated hydrocarbons, which are relatively poor electron donors but which can accept pi electrons from filled *d* orbitals of the metal, tend to stabilize the lowest oxidation states of metals. Intermediate oxidation states are most effectively stabilized by ligands such as water, ammonia, and cyanide ion, which are moderately good sigma electron donors but relatively poor pi electron donors or acceptors (see above *Coordinate bonding*). These trends are illustrated in Table 56, in which the complexes formed by chromium with various ligands are related to the oxidation states and electron configurations.

Table 56: Chromium Complexes of Various Oxidation States

oxidation state	electron configuration*	coordination complex
+6	d^0	$[CrF_6]$, $[CrO_4]^{2-}$
+5	d^1	$[CrO_4]^{3-}$
+4	d^2	$[CrO_4]^{4-}$, $[Cr(OR)_4]$†
+3	d^3	$[Cr(H_2O)_6]^{3+}$, $[Cr(NH_3)_6]^{3+}$
+2	d^4	$[Cr(H_2O)_6]^{2+}$
0	d^6	$[Cr(CO)_6]$, $[Cr(C_6H_6)_2]$

*Number of *d* electrons indicated by superscript.
†Capital R symbolizes an organic alkyl radical.

Aquo complexes. There are few ligands that equal water in respect to the number and variety of metal ions with which they form complexes. Nearly all metallic elements form aquo complexes, frequently in more than one oxidation state. Such aquo complexes include hydrated ions in aqueous solution as well as hydrated salts such as hexaaquochromium(III) chloride, $[Cr(H_2O)_6]Cl_3$. For metal ions with partially filled *d* subshells (*i.e.*, transition metals), the coordination numbers and geometries of the hydrated ions in solution can be inferred from their light-absorption spectra, which are generally consistent with octahedral coordination by six water molecules. Higher coordination numbers probably occur for the hydrated rare-earth ions such as lanthanum(III).

Complex ions in aqueous solution

When other ligands are added to an aqueous solution of a metal ion, replacement of water molecules in the coordination shell may occur, with the resultant formation of other complexes. Such replacement is generally a stepwise process, as illustrated by the following series of reactions that results from the progressive addition of ammonia to an aqueous solution of nickel(II) salt:

$$[Ni(H_2O)_6]^{2+} + NH_3 \rightleftharpoons [Ni(H_2O)_5(NH_3)]^{2+} + H_2O$$

$$[Ni(H_2O)(NH_3)_5]^{2+} + NH_3 \rightleftharpoons [Ni(NH_3)_6]^{2+} + H_2O$$

With increasing additions of ammonia, the equilibria are shifted toward the higher ammine complexes (those with more ammonia and less water) until ultimately the hexaamminenickel(II) ion predominates.

The tendency of metal ions in aqueous solution to form complexes with ammonia as well as with organic amines (derivatives of ammonia, carrying chains of carbon atoms attached to the nitrogen atom) is widespread. The stabilities of such complexes exhibit a considerable range of dependence on the nature of the metal ion as well as on that of the amine. The marked enhancement of stability that results from chelation is reflected in the equilibrium constants of the reactions—values that indicate the relative proportions of the starting materials and the products at equilibrium. Table 57, for example, shows the equilibrium constants for the formation of complexes of hexaaquonickel(II) ions with a series of polyamines—*i.e.,* for a series of reactions of the type

Equilibrium constants

Table 57: Equilibrium Constants for the Formation of Various Nickel Amine Complexes

n	amine (L)	equilibrium constant K_L, M^{-1}*
1	$\underline{N}H_3$	5×10^2
2	$\underline{N}H_2CH_2CH_2\underline{N}H_2$	4×10^7
3	$\underline{N}H_2CH_2CH_2\underline{N}HCH_2CH_2\underline{N}H_2$	5×10^{10}
4	$\underline{N}H_2CH_2CH_2\underline{N}HCH_2CH_2\underline{N}HCH_2CH_2\underline{N}H_2$	1×10^{14}

*M is molar concentrations.

$$[Ni(H_2O)_6]^{2+} + nL \rightleftharpoons [Ni(H_2O)_{6-n}L_n]^{2+} + nH_2O,$$

in which L is the ligand and n the number of water molecules displaced from the complex. In the table it can be seen that the equilibrium constants, K_L, increase dramatically as the possibilities for chelation increase—that is, as the number of nitrogen atoms (N) available for bonding to the metal atom increases.

It should be noted that in the particular examples cited above the coordination number of the metal ion is invariant throughout the substitution process, but this is not always the case. Thus, the ultimate products of the addition of the cyanide ion to an aqueous solution of hexaaquonickel(II) ion are tetracyanonickelate(II) and pentacyanonickelate (II). Similarly, addition of the chloride ion to a solution of hexaaquoiron(III) yields tetrachloroferrate(III).

Halide complexes. Probably the most widespread class of complexes involving anionic ligands is that of the complexes of the halide ions—*i.e.,* the fluoride, chloride, bromide, and iodide ions. In addition to forming simple halide salts, such as sodium chloride and nickel difluoride (in which the metal ions are surrounded by halide ions, these in a sense being regarded as coordinated to them), many metals form complex halide salts, such as potassium tetrachloroplatinate(II), $K_2[PtCl_4]$, that contain discrete complex ions. Most metal ions also form halide complexes in aqueous solution. The stabilities of such complexes span an enormous range—from the alkali-metal ions (lithium, sodium, potassium, and so on), the formation of whose halide complexes in aqueous solution can barely be detected, to extremely stable halide complexes, such as the tetraiodomercurate(II), tetrachlorothallate(III), and tetrachloropalladate(II) ions, the extent of whose dissociation is extremely small.

Hard and soft ions

The stabilities of halide complexes reflect a pattern in which metal ions can be divided into two general classes, designated as A and B or, alternatively, as hard and soft, respectively. (Generally, the electrons in the atoms of the hard elements are considered to form a compact and not easily deformable group, whereas those in the atoms of the soft elements form a looser group—that is, one more easily deformed.) For the former class, which includes beryllium, magnesium, scandium, chromium, iron, nickel, copper, indium, and tin, the order of increasing stability of the halide complexes in aqueous solution is iodides, bromides, chlorides, and fluorides. Conversely, for the class B (or soft) ions, such as platinum, silver, cadmium, mercury, thallium, and lead, the order of increasing stability of the halide complexes is fluorides, chlorides, bromides, and iodides. In contrast to class A metals, those of class B also tend to form more stable complexes with sulfur-containing ligands than with oxygen-containing ligands and more stable complexes with phosphorus ligands than with nitrogen ligands.

Metal carbonyls. Following the discovery of the first metal carbonyl complex, tetracarbonylnickel, in 1890, many compounds containing carbon monoxide coordinated to transition metals have been prepared and characterized. For reasons already discussed, such compounds generally contain metal atoms or ions in low oxidation states. The following are some of the more common types of metal carbonyl compounds: (1) simple mononuclear carbonyls of zero-valent metals, such as tetracarbonylnickel, pentacarbonyliron, and hexacarbonylchromium—highly toxic, volatile compounds, the most stable of which have filled valence shells of 18 electrons; (2) salts of anionic and cationic carbonyls, such as tetracarbonylcobaltate(I) and hexacarbonylmanganese(I); (3) binuclear and polynuclear carbonyls, such as bis(tetracarbonylcobalt), the structural formula of which was shown earlier (see above *Classification: Polynuclear*); and (4) mixed complexes containing other ligands additional to CO: chloropentacarbonylmanganese, hydridotetracarbonylcobalt(I), and tricarbonylnitrosylcobalt.

Complexes of molecular nitrogen

Although molecular nitrogen, N_2, is isoelectronic with carbon monoxide (that is, has the same arrangement of electrons), its tendency to form complexes with metals is much smaller. The first complex containing molecular nitrogen as a ligand—*i.e.,* pentaamminenitrogenruthenium(II), $[Ru(NH_3)_5(N_2)]^{2+}$—was prepared in 1965, and many others have been discovered subsequently. Such complexes have attracted considerable interest because of their possible roles in the chemical and biological fixation of nitrogen.

Organometallic complexes. Organometallic complexes, complexes formed between organic (carbon-containing) groups and metal atoms, can be divided into two general classes: (1) complexes containing metal–carbon sigma bonds and (2) pi-bonded metal complexes of unsaturated hydrocarbons—that is, compounds with multiple bonds between carbon atoms. These substances are covered above in *Organometallic compounds.*

GENERAL METHODS OF PREPARATION

The great variety of coordination compounds is matched by the diversity of methods through which such compounds can be synthesized. Complex halides, for example, may be prepared by direct combination of two halide salts (either in the molten state or in a suitable solvent). Palladium chloride and potassium chloride, for example, react to give the complex potassium tetrachloropalladate(II), as shown in the following equation:

$$PdCl_2 + 2KCl \longrightarrow K_2[PdCl_4].$$

Another widely used route to coordination compounds is through the direct combination of a metal ion and appropriate ligands in solution. Thus, the addition of a sufficiently high concentration of ammonia to an aqueous solution of a nickel(II) salt leads, through a series of reactions (see above *Aquo complexes*) to the formation of the hexaamminenickel(II) ion, which can be precipitated, for example, as the sulfate salt, $[Ni(NH_3)_6]SO_4$.

Complexes of metal ions in high oxidation states are sometimes more readily formed by adding the ligands to a solution of the metal ion in a lower oxidation state in the presence of an oxidizing agent. Thus, addition of ammonia to an aqueous solution of a cobalt(II) salt in the presence of air or oxygen leads to the formation of cobalt(III) ammine complexes such as hexaamminecobalt(III), $[Co(NH_3)_6]^{3+}$, and aquopentaamminecobalt(III), $[Co(NH_3)_5(H_2O)]^{3+}$, ions.

Complexes of metals in low oxidation states, such as the carbonyls of zero-valent metals, can sometimes be prepared by direct combination of the metal with the ligand as, for example, in the reaction of nickel metal with carbon monoxide:

$$\mathrm{Ni(solid)} + 4CO \xrightarrow[1\ \mathrm{atm}]{25^\circ} Ni(CO)_4.$$

More commonly, a salt of the metal is reduced in the presence of the ligand. An example of this type of synthesis is the reduction of cobalt carbonate with hydrogen in the presence of carbon monoxide to give bis(tetracarbonylcobalt):

$$2CoCO_3 + 2H_2 + 8CO \xrightarrow[250-300\ \mathrm{atm}]{120-200^\circ} Co_2(CO)_8 + 2CO_2 + 2H_2O.$$

Similar procedures are applicable to the synthesis of metal sandwich compounds containing cyclopentadienyl and benzene ligands. Dibenzenechromium, for example, can be prepared from chromic chloride, benzene, and aluminum, as shown in the following equation:

$$CrCl_3 + 2C_6H_6 + Al \longrightarrow Cr(C_6H_6)_2 + AlCl_3.$$

Hydrido complexes of transition metals can be prepared by reactions of suitable precursors either with molecular hydrogen or with suitable reducing agents such as hydrazine or sodium borohydride; for example,

$$2[Co(CN)_5]^{3-} + H_2 \longrightarrow 2[CoH(CN)_5]^{3-}$$

Transition metal complexes containing metal–carbon bonds can be prepared by a variety of routes, some of the more important of which are illustrated by the following examples:

$$Mn(CO)_5Br + LiCH_3 \longrightarrow Mn(CO)_5(CH_3) + LiBr,$$

$$[Mn(CO)_5]^- + CH_3Br \longrightarrow Mn(CO)_5(CH_3) + Br^-,$$

$$PtHCl[P(CH_3CH_2)_3]_2 + CH_2{=}CH_2 \longrightarrow Pt(CH_2CH_3)Cl[P(CH_3CH_2)_3]_2,\ \text{and}$$

$$2[Co(CN)_5]^{3-} + CH_3I \longrightarrow [Co(CN)_5(CH_3)]^{3-} + [Co(CN)_5I]^{3-}$$

PRINCIPAL REACTIONS

Acid–base. Coordination to a positive metal ion usually enhances the acidity (*i.e.*, the tendency to lose protons) of hydrogen-containing ligands, such as water and ammonia. Thus, metal ions in aqueous solution commonly exhibit acidic behaviour. Such behaviour is exemplified by hydrolysis reactions of the type shown in the following equilibrium:

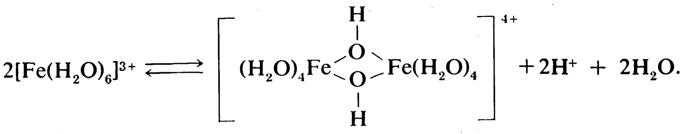
$$[M(H_2O)_x]^{n+} \rightleftharpoons [M(H_2O)_{x-1}(OH)]^{(n-1)+} + H^+,$$

in which M represents the metal ion, n its charge, and x the number of coordinated water molecules.

Acid–base equilibria

The acidities of such aquo ions depend upon the charge, size, and electronic configuration of the metal ion. This dependence is reflected in the values of acid dissociation constants, which range from about 10^{-14} (a value only slightly larger than for pure water, for which the dissociation constant = $10^{-15.7}$) for the hydrated lithium ion, to about 10^{-2} (a value equivalent to that of a fairly strong acid) for the hydrated uranium(IV) ion. Acid–base equilibria are rapidly established in solution, generally within a fraction of a second (see CHEMICAL REACTIONS).

In some cases, hydrolysis of a metal ion may be accompanied by polymerization to form dinuclear or polynuclear hydroxo- or oxygen-bridged complexes:



$$2[Fe(H_2O)_6]^{3+} \rightleftharpoons \left[(H_2O)_4Fe\overset{\overset{H}{|}}{\underset{\underset{H}{|}}{\genfrac{}{}{0pt}{}{O}{O}}}Fe(H_2O)_4 \right]^{4+} + 2H^+ + 2H_2O.$$

Even very weakly acidic ligands, such as ammonia, can acquire appreciable acidity through coordination to a metal ion. Thus, the hexaammineplatinum(IV) ion dissociates according to the equilibrium

$$[Pt(NH_3)_6]^{4+} \rightleftharpoons [Pt(NH_3)_5(NH_2)]^{3+} + H^+.$$

Substitution. One of the most general reactions exhibited by coordination compounds is that of substitution, or replacement, of one ligand by another. This process is depicted in a generalized manner by the equation $ML_{x-1}Y + Z \rightarrow ML_{x-1}Z + Y$ for a metal complex of coordination number x. The ligands L, Y, and Z may be chemically similar or different. (Charges have been omitted here for simplicity.)

Replacement of water

A class of substitution reactions that affords the widest possible comparison of different metal ions is the replacement of water in the coordination shells of metal-aquo complexes in aqueous solution. The substitution may be by another water molecule (which can be labelled with the isotope oxygen-18 to permit the reaction to be followed) or by a different ligand, such as the chloride ion. Reactions of both sorts occur as shown below (oxygen-18 is indicated by the symbol $\overset{*}{O}$):

$$[M(H_2O)_x]^{n+} + H_2\overset{*}{O} \longrightarrow [M(H_2O)_{x-1}(H_2\overset{*}{O})]^{n+} + H_2O$$

$$[M(H_2O)_x]^{n+} + Cl^- \longrightarrow [M(H_2O)_{x-1}Cl]^{(n-1)+} + H_2O.$$

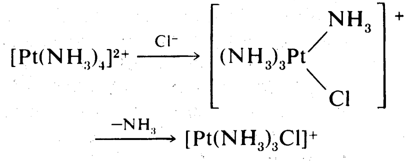
Many such reactions are extremely fast, and it has been only since 1950, following the development of appropriate experimental methods (including stopped flow, nuclear magnetic resonance, and relaxation spectrometry), that the kinetics and mechanisms of this class of reactions have been extensively investigated. Rates of substitution of metal-aquo ions have been found to span a wide range, the characteristic times required for substitution ranging from less than 10^{-9} second for monopositive ions, such as hydrated potassium ions, to several days for certain more highly charged ions, such as hexaaquochromium(III) and hexaaquorhodium(III). The rate of substitution parallels the ease of loss of a water molecule from the coordination shell of the aquo complex and thus increases with increasing size and with decreasing charge of the metal ion. For transition metal ions, electronic factors also have an important influence on rates of substitution.

Dissociative and associative mechanisms

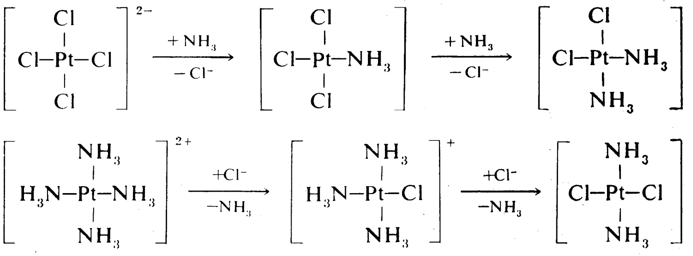
There are two limiting mechanisms (or pathways) through which substitution may occur; namely, dissociative and associative mechanisms. In the dissociative mechanism, a ligand is lost from the complex to give an intermediate compound of lower coordination number. This type of reaction path is typical of octahedral complexes, many aquo complexes, and metal carbonyls such as tetracarbonylnickel. An example of a dissociative reaction pathway for an octahedral complex of cobalt is as follows:

$$[Co(CN)_5(H_2O)]^{2-} \xrightarrow{-H_2O} [Co(CN)_5]^{2-} \xrightarrow{+I^-} [Co(CN)_5I]^{3-}.$$

The associative mechanism for substitution reactions, on the other hand, involves association of an extra ligand with the complex to give an intermediate of higher coordination number; then one of the original ligands is lost to restore the initial coordination number. Substitution reactions of square planar complexes, such as those of the nickel(II), palladium(II), and platinum(II) ions, usually proceed through associative pathways involving intermediates with coordination number 5. An example of a reaction following such a pathway is



A characteristic feature of this class of reactions is the sensitivity of the rate of substitution of a given ligand to the nature of the ligand in the *trans* position (that is, on the opposite side of the metal atom). The *trans* ligand activates a ligand for replacement as follows, in decreasing order: carbon monoxide, cyanide ion, or ethylene, greater than phosphines, or hydride, nitrite, iodide, or thiocyanate ion, greater than bromide or chloride ion, greater than ammonia or water. The *trans* effect may be utilized for synthetic purposes; thus, the reaction of the tetrachloroplatinate(II) ion with ammonia yields *cis*-dichlorodiammineplatinum (II); whereas the reaction of the tetraammineplatinum(II) ion with the chloride ion gives the *trans* isomer, *trans*-dichlorodiammineplatinum(II). The reactions are shown below. In both reactions the *trans* effect causes introduction of the ligand *trans* to chloride rather than *trans* to ammonia:



$$\left[\begin{array}{c}\text{Cl}\\ | \\ \text{Cl–Pt–Cl}\\ | \\ \text{Cl}\end{array}\right]^{2-} \xrightarrow[-\text{Cl}^-]{+\text{NH}_3} \left[\begin{array}{c}\text{Cl}\\ | \\ \text{Cl–Pt–NH}_3\\ | \\ \text{Cl}\end{array}\right] \xrightarrow[-\text{Cl}^-]{+\text{NH}_3} \left[\begin{array}{c}\text{Cl}\\ | \\ \text{Cl–Pt–NH}_3\\ | \\ \text{NH}_3\end{array}\right]$$

$$\left[\begin{array}{c}\text{NH}_3\\ | \\ \text{H}_3\text{N–Pt–NH}_3\\ | \\ \text{NH}_3\end{array}\right]^{2+} \xrightarrow[-\text{NH}_3]{+\text{Cl}^-} \left[\begin{array}{c}\text{NH}_3\\ | \\ \text{H}_3\text{N–Pt–Cl}\\ | \\ \text{NH}_3\end{array}\right]^{+} \xrightarrow[-\text{NH}_3]{+\text{Cl}^-} \left[\begin{array}{c}\text{NH}_3\\ | \\ \text{Cl–Pt–Cl}\\ | \\ \text{NH}_3\end{array}\right]$$

Isomerization. Coordination compounds that exist in two or more isomeric forms (see above *Isomerism*) may undergo reactions that convert one isomer to another. Examples are the linkage isomerization and *cis-trans* isomerization reactions depicted below. The first of these has been shown to proceed intramolecularly (*i.e.*, without dissociation of the nitrite ligand), whereas the second probably occurs through dissociation of one of the water-molecule ligands:

$$[\text{Co(NH}_3)_5(\text{ONO})]^{2+} \longrightarrow [\text{Co(NH}_3)_5(\text{NO}_2)]^{2+}$$

$$cis\text{-}[\text{Co(NH}_2\text{CH}_2\text{CH}_2\text{NH}_2)_2(\text{H}_2\text{O})_2]^{3+} \longrightarrow trans\text{-}[\text{Co(NH}_2\text{CH}_2\text{CH}_2\text{NH}_2)_2(\text{H}_2\text{O})_2]^{3+}.$$

Oxidation–reduction. Transition metals commonly exhibit two or more stable oxidation states, and their complexes accordingly are able to undergo oxidation–reduction reactions. The simplest such reactions involve electron transfer between two complexes, with little if any accompanying rearrangement or chemical change. An example is shown below:

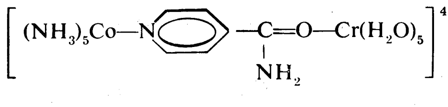
$$[\text{Fe(CN)}_6]^{4-} + [\text{IrCl}_6]^{2-} \longrightarrow [\text{Fe(CN)}_6]^{3-} + [\text{IrCl}_6]^{3-}.$$

In other cases, oxidatinn–reduction is accompanied by significant chemical rearrangement. An example is

$$[\text{Co(NH}_3)_5\text{Cl}]^{2+} + [\text{Cr(H}_2\text{O})_6]^{2+} + 5[\text{H}_3\text{O}]^+ \longrightarrow [\text{Co(H}_2\text{O})_6]^{2+} + [\text{Cr(H}_2\text{O})_5\text{Cl}]^{2+} + 5[\text{NH}_4]^+.$$

Outer-sphere and inner-sphere mechanisms

Two limiting mechanisms of electron transfer, commonly designated outer-sphere and inner-sphere mechanisms, have been recognized. Outer-sphere electron transfer occurs without dissociation or disruption of the coordination shell of either complex; *i.e.*, through both intact coordination shells. The first reaction above is of this type. On the other hand, inner-sphere electron transfer—*e.g.*, the second reaction above—proceeds by formation of a binuclear complex in which the two metal ions are joined by a common bridging ligand (in this case the chloride ion) through which the electron is transferred. Such electron transfer also may occur through polyatomic bridging ligands to which the two metal ions are attached at different sites separated by several atoms; for example, the reduction of pentaammine(isonicotinamide)cobalt(III) by chromium(II) ion through a bridged intermediate,



Strikingly large differences in rates of electron transfer are observed even among closely related reactions. Thus, the rate of reduction of the bromopentaamminecobalt(III) ion by the hexaaquochromium(II) ion is about 10^7 times higher than that of the acetatopentaamminecobalt(III) ion by the same chromium ion.

Oxidative addition. The oxidations of certain complexes, notably those of metal ions with nearly filled *d* shells, are accompanied by increases in their preferred coordination numbers in accord with the patterns shown in Table 55. Such complexes are particularly effective in reducing saturated molecules, such as chlorine, hydrogen peroxide, methyl iodide, and hydrogen, with accompanying incorporation of the fragments of reductive cleavage (the chloride ion, hydroxide ion, methyl carbanion, and hydride ion, respectively) into the coordination shells to achieve the necessary expansions of the coordination numbers. Three such classes of reactions, namely those characteristic of five-coordinate (d^7), four-coordinate (d^8), and two-coordinate (d^{10}) complexes, are illustrated by the equations below:

$$2[\text{Co(CN)}_5]^{3-} + \text{H}_2\text{O}_2 \longrightarrow 2[\text{Co(CN)}_5(\text{OH})]^{3-}$$

$$\text{IrCl(CO)[P(C}_6\text{H}_5)_3]_2 + \text{H}_2 \longrightarrow \text{IrH}_2\text{Cl(CO)[P(C}_6\text{H}_5)_3]_2$$

$$\text{Pt[P(C}_6\text{H}_5)_3]_2 + \text{CH}_3\text{I} \longrightarrow \text{PtI(CH}_3)[\text{P(C}_6\text{H}_5)_3]_2$$

Interest in the synthetic and catalytic applications of such oxidative addition reactions has prompted intensive study of these reactions since about 1960, which has resulted in the discovery of new reactions of this type, and in the elucidation of certain aspects of the mechanisms by which they occur.

SIGNIFICANCE AND APPLICATIONS

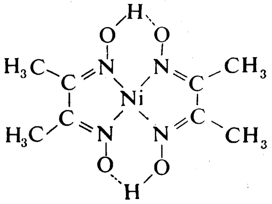
The applications of coordination compounds in chemistry and technology are many and diverse. Naturally occurring coordination compounds are vitally important to living organisms.

Important industrial uses

Dyes and pigments. The brilliant and intense colours of many coordination compounds render them of great value as dyes and pigments. The early use of Prussian blue as a pigment has already been mentioned. Phthalocyanine complexes—for example, copper phthalocyanine—containing large-ring ligands closely related to the porphyrins, constitute an important class of dyes for fabrics.

Extraction and separation of metals. Several important hydrometallurgical processes utilize metal complexes. Nickel, cobalt, and copper can be extracted from their ores as ammine complexes using aqueous ammonia. Differences in the stabilities and solubilities of the ammine complexes can be utilized in selective precipitation procedures that bring about separation of the metals. The purification of nickel can be effected by reaction with carbon monoxide to form the volatile tetracarbonylnickel complex, which can be distilled and thermally decomposed to deposit the pure metal. Aqueous cyanide solutions usually are employed to separate gold from its ores in the form of the very stable dicyanoaurate(I) complex. Cyanide complexes also find application in electroplating.

Chemical analysis. There are a number of ways in which coordination compounds are used in the analysis of various substances. These include (1) the selective precipitation of metal ions as complexes, for example, nickel(II) ion as the dimethylglyoxime complex (shown below); (2) the formation of coloured complexes, such as the tetrachlorocobaltate(II) ion, that can be determined spectrophotometrically—that is, by means of their light absorption properties; and (3) the preparation of complexes, such as metal acetylacetonates, that can be separated from aqueous solution by extraction with organic solvents:



Sequestering. In certain circumstances the presence of metal ions is undesirable, as for example in water, in which calcium and magnesium ions cause hardness. In such cases the undesirable effects of the metal ions frequently can be eliminated by "sequestering" the ions as harmless complexes through the addition of an appropriate complexing reagent. Ethylenediaminetetraacetic acid (EDTA) forms very stable complexes and is widely used for this purpose. Its applications include water softening (by tying up calcium and magnesium ions) and the preservation of organic substances, such as vegetable oils and rubber, in which case it combines with traces of transition metal ions that would catalyze oxidation of the organic substances.

Catalysis. A technological and scientific development of major significance was the discovery in 1954 that certain complex metal catalysts, viz., a combination of titanium trichloride and triethylaluminum bring about the polymerizations (or joining together of the molecules) of organic compounds with carbon–carbon double bonds under mild conditions to form polymers of high molecular weight and highly ordered (stereoregular) structures. Certain of these polymers are of great commercial importance because they are used to make fibres, films, and plastic articles of many kinds. Other technologically important processes based on metal complex catalysts include the catalysis by metal carbonyls, such as hydridotetracarbonylcobalt(I), of the so-called hydroformylation of olefins—*i.e.,* of their reactions with hydrogen and carbon monoxide to form aldehydes—and the catalysis by tetrachloropalladate(II) ions of the oxidation of ethylene in aqueous solution to acetaldehyde (see also CHEMICAL REACTIONS: *Catalysis*).

Biology. Metal complexes play a variety of important roles in biological systems. Many enzymes, the naturally occurring catalysts that regulate biological processes, are metal complexes (metalloenzymes); for example, a hydrolytic enzyme important in digestion, carboxypeptidase, contains a zinc ion coordinated to several amino acid residues of the protein. Another enzyme, catalase, which is a very efficient catalyst for the decomposition of hydrogen peroxide, contains iron–porphyrin complexes. In both cases the coordinated metal ions are probably the sites of catalytic activity. Hemoglobin also contains iron–porphyrin complexes, its role as an oxygen carrier being related to the ability of the iron atoms to coordinate oxygen molecules reversibly. Other biologically important coordination compounds include chlorophyll (a magnesium–porphyrin complex) and vitamin B_{12}, a complex of cobalt with a macrocyclic ligand known as corrin.

(J.Ha.)

ORGANOMETALLIC COMPOUNDS

Organometallic compounds are those substances that contain chemical bonds between carbon atoms and metal atoms; this definition excludes such inorganic salts as the metal carbonates, which show quite different properties. Organometallic compounds constitute an immense group of chemical substances of importance to the history, theory, laboratory practice, and technology of chemistry. In a very general way it may be said that organometallic substances have unique properties that set them apart from both the inorganic and the organic families of compounds.

The metals n organometallic compounds

The metals involved in organometallic compounds may be from any of three categories: (1) the so-called main-group metals of the periodic table of the elements, which include such chemically active metals as lithium and magnesium; (2) the transition series of metals, which include those elements commonly thought of as metallic in character, such as iron, titanium, and platinum; and (3) the metalloids, or partly metal elements, such as silicon and boron. The carbon-containing groups in organometallic compounds also may be of several different varieties, including simple hydrocarbons (groups comprised solely of carbon and hydrogen atoms), unsaturated hydrocarbons (groupings with multiple bonds between carbon atoms), aromatic hydrocarbons (groupings with especially stable multiple bonds), and groups containing atoms other than carbon and hydrogen (oxygen or nitrogen, for example). Metal carbonyls, substances containing metal atoms and units of carbon monoxide, and metal isocyanides, substances containing metals and multiply bonded nitrogen-carbon groups, are classed as organometallic compounds and also as coordination compounds (compounds formed with metal atoms using bonds over and above those normally employed in bonding), revealing the close relationship between organometallic and coordination chemistry.

The carbon–metal bonds in organometallic compounds are of a number of different types: (1) ordinary covalent bonds, or σ-bonds, characterized by pairs of electrons shared between atoms, as, for example, in tetraethyllead, $(C_2H_5)_4Pb$, a gasoline additive; (2) ionic bonds, characterized by complete association of a bonding electron pair with one atom only, resulting in a separation of electrical charge, as in ethylpotassium, $(K^+C_2H_5^-)$; (3) unusual, multicentre covalent bonds, in which bonding extends over a group of atoms (in contrast with the usual two atoms), as in certain polymeric derivatives of light metals, such as ethyllithium, $(C_2H_5Li)_6$; and (4) donor–acceptor bonds, which involve association of a metal atom with an extended system of unsaturation, or overlapping π-orbitals (special electron pathways), as are found in ferrocene, bis(π-cyclopentadienyl) iron, $(\pi\text{-}C_5H_5)_2Fe$.

Physical and chemical properties

A group of compounds as varied as those in the organometallic series naturally includes substances exhibiting a wide range of physical and chemical properties. With regard to stability to heat and to air oxidation they range from alkylpotassium and alkylaluminum compounds, many of which are spontaneously inflammable in air, to ferrocene and the tetraalkyl derivatives of tin, which are very stable. Ferrocene, for example, survives heating to over 470° C (878° F) in the absence of air. The covalent organometallics vary from crystalline solids, often volatile and usually soluble in organic solvents, to liquids—or even, occasionally, to gases at room temperature; for example, trimethylboron, $(CH_3)_3B$, which boils at $-22°$ C ($-7.6°$F). Many organometallic compounds are highly toxic, particularly those that are volatile.

The first references to organometallic compounds appeared in the early 19th century. In 1827, for example, the Danish chemist W.C. Zeise reported that when platinum chloride was boiled with ethanol, and potassium chloride was added to the solution, a salt was obtained having the composition $KCl \cdot PtCl_2 \cdot C_2H_4 \cdot H_2O$. More than 100 years later it was realized that Zeise's salt should be formulated as a coordination compound of platinum, viz., $K[C_2H_4PtCl_3] \cdot H_2O$, with carbon–metal bonding. A reasonable theory to account for the bonding of the ethylene molecule (the C_2H_4 unit) to the platinum ion in this compound was not suggested until 1953.

Early historical landmarks in organometallic chemistry were the synthesis of the arsenic compound cacodyl, $(CH_3)_2AsAs(CH_3)_2$, in 1842 by the German chemist Robert Bunsen; the preparation of alkylzinc compounds in 1849 by the English chemist Sir Edward Frankland; the synthesis of nickel carbonyl in 1890 in England by Ludwig Mond, a German-born chemist; and the discovery of organomagnesium compounds in 1900 by the French chemist Victor Grignard. Although organometallic compounds had been studied for over a century, it was not until 1951, when the highly stable substance called ferrocene was discovered and its unusual π-bonded structure became known, that organometallic compounds in themselves became the subject of truly intensive research. Up to that time organometallic compounds had been of importance chiefly as substances that were useful in the synthesis of organic compounds of various types. This situation was particularly true of the organomagnesium compounds which are called, after their discoverer, Grignard reagents.

BONDING IN ORGANOMETALLIC COMPOUNDS

Polar and nonpolar bonding

Metals are electropositive elements—that is, their atoms tend to lose electrons to the atoms of other elements with which they combine. The atoms of nonmetals, on the other hand, generally attract electrons in chemical compounds. The concept of electronegativity (the power of an atom in a compound to attract electrons to itself) is useful in understanding the nature of chemical bonds. Thus, all metals have electronegativities less than that of carbon, with the result that in every carbon–metal bond the bonding electrons are attracted somewhat more to the carbon atom than to the metal atom, and the bond is polar (shows a separation of electrical charges) to some degree. Differences in electronegativity may be sufficient to produce compounds containing discrete positive and negative ions (atoms that have gained or lost electrons) held together by electrostatic forces. The electronegativity of a carbon atom is enhanced when it is part of an unsaturated system or when halogen atoms, particularly those of fluorine, are joined to it. These charge-transfer, or electrostatic, effects brought about by unsaturation or halogen substitution are responsible for the compound sodium cyclopentadienide having a structure with discrete Na^+ and $C_5H_5^-$ ions, and for the compound bis(trifluoromethyl)mercury, $(CF_3)_2Hg$, resembling more the ionic substance mercuric chloride, $HgCl_2$, than the covalent dimethylmercury, $(CH_3)_2Hg$. The latter substance, unlike its fluorocarbon analogue, is a typically organic, volatile liquid (boiling point 96° C [204.8° F]), immiscible with water. More organometallic compounds are covalent in nature than are truly ionic, because electronegativity differences between the organic groups and the metal atoms are not sufficiently pronounced to do other than impart a degree of polarity to a covalent bond. Indeed, with the least electropositive metals, such as tin or lead, their organic derivatives provide excellent examples of covalent compounds, with classical two-electron σ-bonds (bonds formed from ordinary σ-orbitals, as opposed to those from π-orbitals). Consequently, like hydrocarbons (compounds of carbon and hydrogen), these compounds are unaffected by oxygen and water at room temperature.

Organometallic compounds formed from the lighter metals (for example, lithium, beryllium, or aluminum) show somewhat different behaviour. In these compounds the carbon–metal bonds are readily decomposed by oxygen or water, affording metal–oxygen bonds and releasing the organic groups as hydrocarbons. The high reactivity of these covalent carbon–metal bonds is due to their considerable polarity, coupled with the ability of the metals to coordinate (form "secondary" bonds) with reactant molecules as a prelude to decomposition. Indeed, the availability of vacant metal orbitals to accept electron pairs is reflected in the polymeric nature of many of the organic derivatives of lithium, beryllium, and aluminum—*e.g.*, the dimethylberyllium compound, $[(CH_3)_2Be]_x$; the alkyl groups in these compounds link the metal atoms together, and to account for this it is assumed that the electron pairs occupy bonding orbitals embracing more than two atoms. These compounds, in which the number of bonding orbitals exceeds the number of electron pairs, are termed electron-deficient; they represent a bonding akin to that found in the metals, where the electrons are delocalized over many atoms in the structure.

Organometallic compounds of the transition elements

The transition metal–organic compounds include a number of metal alkyls, such as bis(triphenylphosphine)dimethylplatinum $[(C_6H_5)_3P]_2Pt(CH_3)_2$, which contain covalent carbon–metal σ-bonds similar to those found in organometallic compounds of the main-group elements, such as tin. Unless other groups are simultaneously coordinated to the metal atoms, however, most compounds of this class are of low thermal or oxidative stability. Occupation of all coordination sites (typically six) of a metal atom has an important influence on stability. Vacant sites allow coordination of oxygen and water molecules, circumstances that lead to decomposition by paths requiring low activation energies (energies required to initiate reactions). Thus, dimethylmanganese, $Mn(CH_3)_2$, is a bright-yellow powder that readily detonates and reacts instantaneously with air, whereas pentacarbonylmethylmanganese, $CH_3Mn(CO)_5$, is a volatile white solid, mp 95° C (203° F), stable to air for long periods in the solid state, and decomposing only slowly in solution.

The effective atomic number rule

In addition to the purely σ-complexes, the transition metals also form a large number of coordination compounds in which an unsaturated organic system is associated with the metal. In addition to Zeise's salt and ferrocene (see above), examples include allylpalladium chloride, tricarbonyl (cyclobutadiene) iron, and bis(benzene)chromium. These compounds are called π-complexes because the π-electrons of the hydrocarbon moieties (combining units) are involved in bonding to the metal atoms. The bonding in these cases is thought to involve reciprocal electron transfers between vacant and occupied π-orbitals of the organic group and filled and unfilled orbitals (electron pathways not involved in ordinary valence bonding) of the metal atom. Generally the metals are in their lower formal valence state—that is, fewer of their *d* electrons are involved in direct bonding—because this condition allows an energetically more favourable interaction of the *d* orbitals with the organic moieties. Many organic derivatives of the transition metals obey the so-called effective-atomic-number rule (a method for determining the degree of electronic saturation), with the result that the final electronic configuration of a fully coordinated metal is that of the next noble gas of higher atomic number. Thus, in the complex $\pi\text{-}C_5H_5Fe(CO)_2CH_3$, the iron atom acquires an additional 10 electrons to reach the configuration of the noble gas krypton, which has 18 more electrons than the preceding noble gas, argon; these electrons are supplied by the two CO groups (four electrons), the π-cyclopentadienyl group (five electrons), and the methyl group (one electron).

Many organo-transition metal complexes show fluxional behaviour—that is, the molecules exist in several chemically equivalent structures and are able to pass from one to another rapidly, a process that may be detected by the technique of nuclear magnetic resonance spectroscopy, which measures certain magnetic phenomena associated with the atomic nuclei.

ORGANOMETALLIC COMPOUNDS OF THE MAIN-GROUP ELEMENTS

Organic compounds of the main-group metals are generally named as a combination of the organic group and the metal; *e.g.*, $Si(C_2H_5)_4$, tetraethylsilicon; C_6H_5HgCl, phenylmercuric chloride; and $Na[B(C_6H_5)_4]$, sodium tetraphenylborate. Occasionally, compounds are named by adapting the name of the metal hydride, in the manner of organic compounds; *e.g.*, $(CH_3)_3SnH$ is called trimethylstannane.

Certain regularities in the trends of properties in the various families of organometallic compounds can be observed. For example, within a family, boiling points increase with increasing atomic weight of the metal atom; thus the boiling points for the tetramethyl derivatives of Group IVa elements (other than carbon) are as follows: silicon, 26.5° C (79.7° F); germanium, 43.4° C (110° F); tin, 78° C (172.4° F); and lead, 110° C (230° F). Chemical reactivities of organometallic compounds also vary in regular patterns. The small size of the lithium atom causes organolithium compounds to have covalent properties, but the other Group Ia organometallics show a steadily increasing reactivity from sodium to cesium.

Methods of preparation. The formation of these compounds may be achieved by three main methods: (1) reaction of a free metal with an organic halide or an unsaturated organic compound; (2) displacement of one metal in an organometallic compound by another, by reaction with a halide of the second metal; and (3) the addition of metal hydrides to unsaturated organic compounds.

1. Reactive main-group metals, such as lithium, sodium, and magnesium, react with organic halides to give either the organometallic compound or an organometallic halide. Examples are:

$$n\text{-}C_4H_9Br + 2Li \longrightarrow n\text{-}C_4H_9Li + LiBr$$

$$(C_6H_5)_3CCl + 2Na \longrightarrow (C_6H_5)_3CNa + NaCl$$

$$CH_3I + Mg \longrightarrow CH_3MgI\,.$$

Some hydrocarbons—*e.g.*, acetylene and cyclopentadiene—are sufficiently acidic to react directly with a metal, the reaction being accompanied by evolution of hydrogen:

$$C_6H_5C{\equiv}CH + Na \longrightarrow C_6H_5C{\equiv}CNa + \tfrac{1}{2}H_2$$

$$C_5H_6 + K \longrightarrow C_5H_5K + \tfrac{1}{2}H_2\,.$$

2. Treatment of a metal halide with an organic derivative of a different metal often leads to the substitution of one metal for the other. This is the most common synthetic method, and it is one of wide scope. A variety of examples are shown:

$$3(C_2H_5)_3Al + GaBr_3 \longrightarrow (C_2H_5)_3Ga + 3(C_2H_5)_2AlBr$$

$$(C_6H_5)_4Sn + 2BCl_3 \longrightarrow 2C_6H_5BCl_2 + (C_6H_5)_2SnCl_2$$

3. The readily available hydrides of metals of Groups IIIa and IVa add across carbon–carbon double and triple bonds, on heating or on exposure to ultraviolet light, often in the presence of a catalyst:

$$(C_6H_5)_3SnH + CH_2{=}CHCN \longrightarrow (C_6H_5)_3SnCH_2CH_2CN$$

$$Cl_3SiH + CH{\equiv}CH \longrightarrow Cl_3SiCH{=}CH_2$$

$$6n\text{-}C_4H_9CH{=}CH_2 + B_2H_6 \longrightarrow 2(C_6H_{13})_3B\,.$$

In an extensively used industrial process, olefins with terminal methylene, $=CH_2$, groups (1-olefins) react with aluminum and hydrogen under pressure to give the alkyls directly:

$$Al + \tfrac{3}{2}H_2 + 3(\text{1-olefin}) \longrightarrow (\text{alkyl})_3Al.$$

Specific synthetic procedures

The syntheses of many specific organometallics are dictated by the particular properties or reactions of a single compound or group of compounds. Thus, cyclopentadienylthallium is insoluble in water, and it may be prepared from cyclopentadiene and thallium hydroxide in aqueous solution because it readily separates from the solution. A large number of organomercury compounds are available by unique reactions: (1) mercuration of aromatic compounds with mercuric acetate, (2) reaction of aryldiazonium salts with mercury, and (3) interaction of arylsulfinic acids with mercuric chloride:

$$C_6H_6 + Hg(O_2CCH_3)_2 \longrightarrow C_6H_5HgO_2CCH_3 + CH_3CO_2H$$

$$ArN_2Cl + Hg \longrightarrow ArHgCl + N_2$$

$$ArSO_2H + HgCl_2 \longrightarrow ArHgCl + SO_2 + HCl,$$

where Ar is an aryl group.

The organosilicon halides, which are technologically important because they are precursors to the silicones, are prepared by a special reaction using copper as catalyst:

$$2CH_3Cl + Si \xrightarrow[300°]{Cu} (CH_3)_2SiCl_2$$

Organogermanium and organotin halides may be prepared similarly.

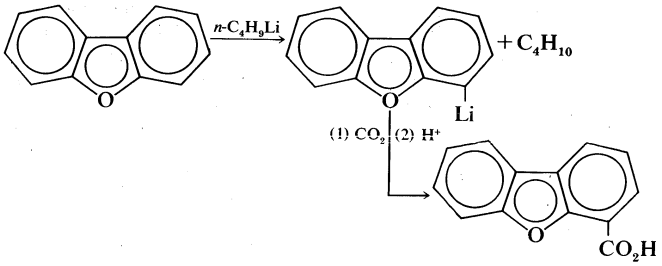
Many organoboron compounds have compact molecular structures based on cages of boron atoms and carbon atoms ("carboranes"). The most easily prepared carborane is $C_2B_{10}H_{12}$, which has a structure based on an icosahedron (a polyhedron with 20 faces), with 10 boron and two adjacent carbon atoms at the apexes and one hydrogen atom attached to each of these. This carborane can be prepared as follows:

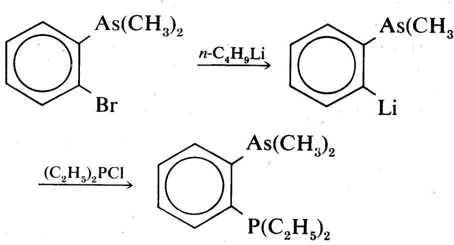
$$B_{10}H_{14} + HC{\equiv}CH \xrightarrow[C_6H_6]{2(C_2H_5)_2S} C_2B_{10}H_{12} + 2H_2.$$

The compound is high melting (294.5–295.5° C [562.1–563.9° F]) and is very resistant to oxidation and hydrolysis. Substitution at the carbon atoms may be effected by treatment with butyllithium to give $Li_2C_2B_{10}H_{10}$, followed by reaction with reactive halides.

Principal reactions. The reactions of main-group organometallic compounds are extremely varied, and certain classes of compound—*e.g.*, lithium, boron, and aluminum compounds, and the Grignard reagents—are very useful synthetically.

Organolithium compounds often are not isolated but are used as soon as they are formed, and a frequently employed sequence of reactions is metalation of an organic compound with *n*-butyllithium (a commercially available product) followed by reaction with a selected reagent that reacts with the organometallic intermediate.





A useful reaction occurs with alkyltriphenylphosphonium salts to give phosphinemethylenes. These intermediates react with carbonyl compounds to give olefins, a reaction known as the Wittig reaction.

$$(C_6H_5)_3P \xrightarrow{CH_3Br} (C_6H_5)_3P(CH_3)^+Br^- \xrightarrow{C_6H_5Li}$$

$$(C_6H_5)P = CH_2 \xrightarrow{R_2CO} (C_6H_5)_3PO + R_2C = CH_2$$

The most important reactions of organoboron compounds, R_3B, are cleavage of the C–B bonds in various ways to give alkanes, alcohols, ketones, and other compounds. The addition of B–H bonds to carbon–carbon double bonds is called hydroboration; the reaction proceeds rapidly in ethers (as solvents) affording diverse species, R_3B, in which the R groups can be complex organic moieties. Many syntheses of purely organic compounds depend on the addition of diborane (B_2H_6), often generated *in situ,* to olefins using a polyether—*e.g.*, diglyme, $(CH_3OCH_2CH_2)_2O$—as solvent.

Hydroboration

$$12RCH{=}CH_2 + 3NaBH_4 + 4(C_2H_5)_2OBF_3 \longrightarrow 4(RCH_2CH_2)_3B + 3NaBF_4 + 4(C_2H_5)_2O$$

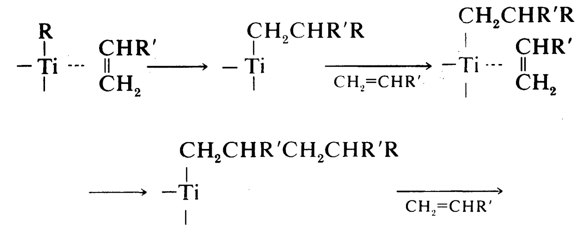
The resulting organoboron compounds undergo a variety of reactions. For example, treatment with alkaline hydrogen peroxide yields an alcohol RCH_2CH_2OH, whereas protonation (addition of a hydrogen ion) gives RCH_2CH_3; *i.e.*, effects hydrogenation of the starting olefin. Oxidation of $(RCH_2CH_2)_3B$ with chromic acid gives ketones or acids, while treatment with diethylchloramine, $(C_2H_5)_2NCl$, gives RCH_2CH_2Cl (corresponding to an overall addition of hydrogen chloride to the olefin in a manner opposite to that normally observed). Thus, a family of related organic syntheses depends on hydroboration.

The industrial availability of triorganoaluminum compounds has resulted in their becoming reagents of great utility. In the preparation of alkyl derivatives of other metals their use avoids the employment of ethers, which are necessary with Grignard reagents. Aluminum alkyls also cause dimerization, oligomerization, or polymerization of olefins—that is, joining of two olefin molecules, several olefin molecules, or many, respectively. The reactions are of use in the manufacture of detergents, which often are made from long-chain alcohols, as follows:

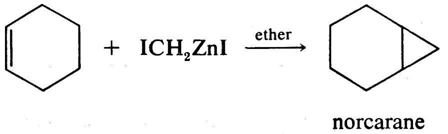
$$(C_2H_5)_3Al \xrightarrow[\substack{\sim 100^\circ \\ \sim 100\ \text{atm}}]{C_2H_4} [C_2H_5(C_2H_4)_n]_3Al \xrightarrow[\text{oxidation}]{\text{controlled}} H(C_2H_4)_{n+1}OH.$$

Olefin polymerization

If a compound $[CH_3(CH_2)_n]_3$ Al is heated under pressure, the alkyl chain becomes detached in the form of a medium-to-high-molecular-weight linear polyethylene. Such polymers, however, are better obtained by using aluminum–transition metal catalysts (Ziegler–Natta catalysts). For example, treatment of titanium tetrachloride, ($TiCl_4$), with triethylaluminum $(C_2H_5)_3Al$, in heptane solvent gives a complex that reacts with such olefins as ethylene or propylene. The polymers produced by this reaction are isotactic (sterically regular) solids with desirable properties, as opposed to disordered (atactic) polymers that are oils at room temperature. The mechanism of this industrially important process is a matter of some controversy but probably involves an initial reduction and alkylation of the titanium by the aluminum alkyl, followed by coordination and insertion of olefin molecules in a stereoregular way.



A mixture of zinc and copper reacts with di-iodomethane to give iodomethylzinc iodide (stable in solution only). This compound is useful in reacting with olefins to give cyclopropanes:



norcarane
(bicyclo [4.1.0] heptane)

Organozinc compounds also are formed as transient intermediates in a standard reaction for producing highly substituted esters (the Reformatsky reaction):

$$BrCH_2CO_2C_2H_5 \xrightarrow{Zn} [BrZnCH_2CO_2C_2H_5] \xrightarrow[H_2O]{R'_2CO} R'_2C(OH)CH_2CO_2C_2H_5.$$

The most important reaction of organocadmium compounds is the reaction with acid chlorides to give ketones:

$$2RCOCl + R'_2Cd \longrightarrow 2RCOR' + CdCl_2.$$

This method is preferred to the use of Grignard reagents, since the latter react rapidly with the ketone produced.

Analytical chemistry. Organic derivatives of main-group metals may be analyzed by combustion analysis, in which the compound is burned and the carbon and hydrogen present are determined as carbon dioxide and water (the residual metal oxide may be weighed directly). Spectroscopic methods—*e.g.*, infrared and nuclear magnetic resonance spectroscopy—provide a great deal of information, especially about the nature of the organic groups present. Mass spectrometry also is extensively used, often yielding precise molecular-weight measurements.

Tetraethyllead

Industrial applications. Aluminum alkyls are employed industrially in olefin polymerization and in the synthesis of long-chain alcohols (see above). Tetraethyllead was formerly used on a large scale as an additive (antiknock agent) in gasoline, since it is a combustion catalyst that permits higher compression ratios in the engine. The commercial process for manufacture of tetraethyllead is as follows:

$$4NaPb + 4C_2H_5Cl \longrightarrow (C_2H_5)_4Pb + 4NaCl + 3Pb.$$

In gasoline, tetraethyllead is mixed with dibromoethane and dichloroethane; these substances act as "scavengers" within the engine by converting the resulting lead oxide into volatile halides, which are emitted in the exhaust. Volatile lead compounds, however, are highly toxic, and their expulsion into the atmosphere is undesirable. Hence the use of lead alkyls in motor fuels is declining, and other methods are being sought for the production of fuels with superior combustion characteristics (higher "octane number").

Controlled hydrolysis of organosilicon halides produces polymers containing silicon–oxygen and silicon–carbon bonds. The silicones have unusual properties and are widely used for a number of purposes. Organotin compounds are used to stabilize polyvinyl chloride, a polymer used in many common plastic articles. They are also used as fungicides in agriculture and in treating plant diseases that infect potatoes, sugar beets, cacao, groundnuts (peanuts), and other crops. Tricyclohexyltin hydroxide also is used in agriculture, to kill mites. Tributyltin oxide is important as an industrial biocide and surface disinfectant; thus, it is used in preserving wood and in keeping clean the recirculating water of cooling towers.

Certain organomercury compounds have been widely used in pharmacy, for example, as diuretics, but have now been largely replaced by more specific organic compounds. Mercurochrome (2,7-dibromo-4-hydroxymercurifluorescein) is a familiar antiseptic.

ORGANOMETALLIC COMPOUNDS OF THE TRANSITION ELEMENTS

Organic compounds of transition metals are generally named as a combination of the organic group and the metal; *e.g.*, $Mn(CH_3)_2$, dimethylmanganese. When other donor ligands are present, their names precede that of the organic group; *e.g.*, $[(C_6H_5)_3P]_2Pt(C_6H_5)I$, iodobis(triphenylphosphine)phenylplatinum. Similarly, the names of π-bonded organic groups precede those of σ-bonded groups; *e.g.*, $(\pi\text{-}C_5H_5)Fe(CO)_2CH_3$, dicarbonyl-$\pi$-cyclopentadienyl-$\sigma$-methyliron. As these names tend to be cumbersome, many of the commonly encountered compounds have acquired simpler names; *e.g.*, $Fe(\pi\text{-}C_5H_5)_2$, ferrocene or bis(π-cyclopentadienyl)iron; $(\pi\text{-}C_5H_5)Mn(CO)_3$, cymantrene or π-cyclopentadienyltricarbonylmanganese. Organic complexes of the transition elements may be classified according to the number of electrons donated by the organic ligand (that is, the substance complexed with the metal atom; see Table 58). A great many separate compounds are known because many members of each particular class of organic molecule can bond to a given metal. There are, for example, a vast number of complexes of formula diene-$Fe(CO)_3$ because the $Fe(CO)_3$ group shows great affinity for any 4π-electron system based on two conjugated double bonds, of which there are many. Moreover, a particular metal may form complexes with all types of ligand; for example, in the series $CH_3Mn(CO)_5$,

$$[(\pi\text{-}C_2H_4)Mn(CO)_5]^+,\ (\pi\text{-}C_3H_5)Mn(CO)_4,$$
$$(\pi\text{-}C_6H_8)Mn(CO)_3H,\ (\pi\text{-}C_5H_5)Mn(CO)_3,\text{ and}$$

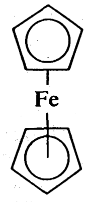
$[(\pi\text{-}C_6H_6)Mn(CO)_3]^+$, manganese is associated with organic groups that donate—in order—from one to six electrons to its valence shell. In each case the effective atomic number rule is obeyed, the manganese atom having 18 electrons in its outermost shell.

Table 58: Representative Ligands in Organotransition-Metal Complexes

number of electrons	class	examples of complexes
1	yl	$CH_3Re(CO)_5$, white, mp 120°
2	alkene	$[(\pi\text{-}C_2H_4)_2RhCl]_2$, orange red, decomp. > 100°
	carbene	$CH_3(CH_3O)CCr(CO)_5$, white, mp 34°
3	enyl*	$[(\pi\text{-}C_3H_5)PdCl]_2$, yellow, mp 145°
4	diene	$(\pi\text{-}C_4H_6)Fe(CO)_3$,† yellow oil
5	dienyl	$(\pi\text{-}C_5H_5)_2Ni$, green, mp 173°
	ollyl	$[(CH_3)_4N][\{\pi\text{-}1,2\text{-}B_9H_9C_2(CH_3)_2\}_2Fe]$‡, red, mp 247°
6	arene	$(\pi\text{-}C_6H_6)_2W$, green, decomposes 110°
	triene	$(\pi\text{-}C_7H_8)Mo(CO)_3$,§ orange red, mp 101°
7	trienyl	$(\pi\text{-}C_7H_7)V(CO)_3$, dark green, decomposes 134°

*Also called "π-allylic" ligands. †C_4H_6 is buta-1,3-diene. ‡The ligand $\{1,2\text{-}B_9H_9C_2(CH_3)_2\}^{2-}$ is an 11-atom B_9C_2 species (an icosahedron with an apex removed) having a face comprising five atoms (B_3C_2) with π-type molecular orbitals similar to those of $\pi\text{-}C_5H_5$. In forming the complex the metal atom completes the icosahedron. §C_7H_8 is cycloheptatriene.

Although this rule is very useful, it is not followed invariably; the exceptions may be rationalized by considering the relative energies of the *s, p,* and *d* orbitals (designations for electron pathways within the various shells). For metals of the scandium and titanium groups of the periodic table, the 3*d* orbitals have a relatively high energy compared with the other orbitals (4*s* and 4*p*) in the same valence band, and their importance in bonding, therefore, is less. At the other end of the transition series (in the neighbourhood of copper and zinc), the *d* orbitals are of relatively low energy and cannot easily be regarded as valence orbitals. Across the series vanadium, chromium, manganese, iron, cobalt, and nickel, the 3*d* orbitals decrease in energy more rapidly than do the 4*s* and 4*p* orbitals, but all three orbital types are generally sufficiently close in energy to be suitable for metal–ligand bonding. Under these conditions, maximum bonding—with associated stability—occurs when all orbitals are filled with electron pairs. This occurs when the metal atom has an 18-electron configuration; *i.e.,* the five *nd,* one $(n+1)s$, and three $(n+1)p$ orbitals ($n = 3$, 4, or 5) are all occupied by metal and ligand electrons. Nowhere is this better exemplified than in the case of ferrocene, a compound of exceptional stability and one in which the iron atom (eight electrons) is located symmetrically between two $\pi\text{-}C_5H_5$ groups (ten electrons), forming a classical "sandwich" molecular structure.



Exceptions to the effective-atomic-number rule

It follows from this reasoning that exceptions to the effective-atomic-number rule are found at both ends of the transition series, in which cases the 3*d* orbitals are too high (as with titanium) or too low (copper) in energy. Sixteen-electron environments are prevalent with organotitanium compounds—*e.g.,* $(\pi\text{-}C_5H_5)_2Ti(C_6H_5)_2$—and 14 with the metals of the copper group [*e.g.,* $(C_6H_5)_3PAuCH_3$]. With the latter, not only have the *d* electrons become low-energy (core) electrons but the *p* orbitals have become relatively high in energy compared with the *s* orbitals; consequently only one *p* orbital is used in bonding, in combination with the *s* orbital. A similar effect is responsible for the existence of numerous 16-electron complexes of nickel, palladium, and platinum, in which the relative energies of the *nd,* the $(n+1)s$, and the $(n+1)p$ orbitals are such as to lead to the use of one of the *d,* the *s,* and two of the *p* orbitals in bonding—*e.g.,* $[(C_2H_5)_3P]_2Pd(CH_3)I$.

Methods of preparation. It is convenient to consider separately the methods for preparing the π-complexes and the σ-compounds of the transition metals.

π-Complexes. Treatment of a suitable metal halide with an organic derivative of an alkali metal or a Grignard reagent using an ether (often tetrahydrofuran) as solvent is often a method for preparing π-complexes of transition elements.

$$2CH_2{=}CHCH_2MgBr + NiCl_2 \longrightarrow (\pi\text{-}C_3H_5)_2Ni + 2MgClBr$$

$$2C_5H_5Na + FeCl_2 \longrightarrow (\pi\text{-}C_5H_5)_2Fe + 2NaCl$$

$$2C_8H_8K_2 + UCl_4 \longrightarrow (\pi\text{-}C_8H_8)_2U + 4KCl$$

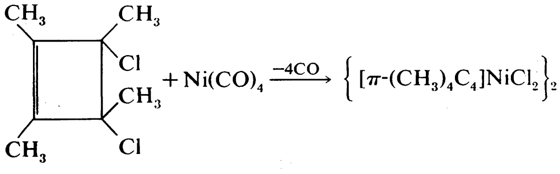
Reactions with olefins

The reaction between a metal complex and an unsaturated hydrocarbon is the most generally useful method of preparing π-complexes; it has several modifications. In one of the simplest forms of this reaction, carbonyls are displaced from metal carbonyl complexes by unsaturated hydrocarbons. The reaction is carried out by heating the reactants together or by exposing them to ultraviolet irradiation in a suitable solvent (often a saturated hydrocarbon).

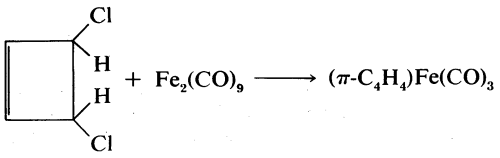
$$Fe(CO)_5 + CH_2{=}CHCH{=}CH_2 \xrightarrow{-2CO} (\pi\text{-}C_4H_6)Fe(CO)_3$$

$$CO_2(CO)_8 + 2C_5H_6 \longrightarrow 2(\pi\text{-}C_5H_5)Co(CO)_2 + 4CO + H_2$$

The more reactive metal carbonyls abstract chlorine atoms from carbon–chlorine bonds with concomitant oxidation of the metal to a metal chloride. This property has been used to isolate organic molecules that are not normally stable in the free state but that can be stabilized by coordination with a metal atom. Cyclobutadiene and its derivatives, sought by chemists for over 80 years, were finally prepared by being stabilized in this way.



$$+ Ni(CO)_4 \xrightarrow{-4CO} \left\{[\pi\text{-}(CH_3)_4C_4]NiCl_2\right\}_2$$



$$+ Fe_2(CO)_9 \longrightarrow (\pi\text{-}C_4H_4)Fe(CO)_3$$

Another example of this type of stabilization is found in the preparation of the trimethylenemethane group:

$$CH_2{=}C(CH_2Cl)_2 + Fe_2(CO)_9 \longrightarrow [\pi\text{-}(CH_2)_3C]Fe(CO)_3.$$

Reactions with acetylenes

Reactions between acetylenes and metal carbonyls (or their derivatives) can be a fruitful source of unusual organometallic compounds. A variety of reactions is possible, including coordination of the triple bond, di- or trimerization of the acetylene, and formation of ligands containing carbonyl groups, with the result that highly complex mixtures of products are obtained. For example, the mixture that results from treatment of iron carbonyls with acetylenes has given complexes of nearly 30 different types, containing such diverse ligands as cyclobutadienes, cyclopentadienones, and quinones.

The preparation of many π-complexes depends upon an initial reduction of a transition-metal compound in the presence of an unsaturated hydrocarbon to serve as the ligand. The synthesis of bis(benzene)chromium depends upon this technique:

$$3CrCl_3 + 2Al + AlCl_3 + 6C_6H_6 \longrightarrow 3[(\pi\text{-}C_6H_6)_2Cr]^+[AlCl_4]^-$$

$$2[(\pi\text{-}C_6H_6)_2Cr]^+ + S_2O_4^{2-} + 4OH^- \longrightarrow 2(\pi\text{-}C_6H_6)_2Cr + 2H_2O + 2SO_3^{2-}.$$

With this procedure, or modifications of it, arene complexes of many of the transition metals have been prepared. Aromatics other than benzene also may be used.

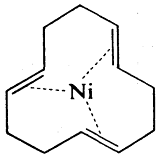
Reduction of nickel acetylacetonate with aluminum alkyls in the presence of triphenylphosphine and ethylene produces a yellow crystalline compound,

$$(\pi\text{-}C_2H_4)Ni[P(C_6H_5)_3]_2.$$

With cycloocta-1,5-diene, rather than ethylene and triphenylphosphine, the same method produces

$$(\pi\text{-}C_8H_{12})_2Ni,$$

a yellow solid of melting point 142° C (287.6° F) that is unstable in air but is reasonably stable to heat. An unusual complex that is prepared by a similar method is all-*trans*-1,5,9-cyclododecatrienenickel,



in which the metal atom is coordinated only to three double bonds. Many olefin complexes of nickel are used in further syntheses. Thus, $(\pi\text{-}C_{12}H_{18})Ni$ catalyzes the trimerization of butadiene, giving isomers of cyclododecatriene. In some syntheses the hydrocarbon itself functions as the reducing agent. Thus, treatment of rhodium trichloride with butadiene gives bis(butadiene)rhodium(I) chloride, $(\pi\text{-}C_4H_6)_2RhCl$.

Ligand-exchange reactions

Ligand-exchange reactions, those in which an organic group is transferred from one metal to another, also are used to prepare π-complexes. For example, treatment of $(C_6H_5CN)_2PdCl_2$ with diphenylacetylene, followed by reaction with hydrogen halides, gives dimeric tetraphenyl-cyclobutadienepalladium halides $\{[\pi\text{-}(C_6H_5)_4C_4]PdX_2\}_2$ (X = Cl or Br). The tetraphenylcyclobutadiene group then may be readily transferred to other metals. Thus, reaction of one of the above palladium complexes with $(\pi\text{-}C_5H_5)_2Co$ gives $[\pi\text{-}(C_6H_5)_4C_4]Co(\pi\text{-}C_5H_5)$, and reaction with $Fe(CO)_5$ gives $[\pi\text{-}(C_6H_5)_4C_4]Fe(CO)_3$.

σ-Compounds. Although simple alkyls and aryls of the transition metals are generally very unstable—*e.g.*, $(CH_3)_2Ni$ decomposes above −110° C (−166° F)—derivatives containing such ligands as carbon monoxide, phosphines, and π-bonded hydrocarbons are often reasonably robust chemically. A number of synthetic routes to these compounds are known.

Treatment of transition-metal-complex halides with Grignard reagents or with organolithium compounds frequently leads to replacement of the halogen with alkyl or aryl groups.

$$(C_6H_5)_3PAuCl + CH_3Li \longrightarrow (C_6H_5)_3PAuCH_3 + LiCl$$

Reactions between complex metal anions and organic halides yield organometallic compounds in which the anion has replaced the halide.

$$Na^+[Mn(CO)_5]^- + CH_3I \longrightarrow CH_3Mn(CO)_5 + NaI$$

The utility of this method is enhanced by the fact that acyl complexes, prepared by reaction of the anions with acid halides, can often be decarbonylated, either thermally or on ultraviolet irradiation, to give the corresponding alkyl compounds:

$$Na^+[Fe(CO)_2(\pi\text{-}C_5H_5)]^- + C_6H_5COCl \longrightarrow C_6H_5COFe(CO)_2(\pi\text{-}C_5H_5) + NaCl \xrightarrow{h\nu} C_6H_5Fe(CO)_2(\pi\text{-}C_5H_5) + CO.$$

Insertion reactions

Insertion of unsaturated groups into metal–hydrogen or metal–alkyl σ-bonds produces organometallic compounds with more extended alkyl chains.

$$(\pi\text{-}C_5H_5)Fe(CO)_2H + CH_2{=}CHCH{=}CH_2 \longrightarrow CH_3CH{=}CHCH_2Fe(CO)_2(\pi\text{-}C_5H_5)$$

A useful synthesis depends on the so-called oxidative-addition or oxidative-elimination reactions. Many transition-metal complexes, in which the metal atom is of low valence and which are not so fully coordinated as they might be, react with electron-acceptor (electrophilic) molecules to form compounds in which the valence state of the metal can be regarded formally as having increased by two units. The reacting molecule, which contains a bond designated A—B, may be cleaved by the metal atom to form in its coordination sphere two separate bonds, M—A and M—B; or, if the bond between A and B is a multiple one, the product may contain an A—B unit coordinated to the metal atom. The following reactions exemplify the utility of these kinds of processes in organometallic syntheses:

$$Fe(CO)_5 + C_3F_7I \longrightarrow C_3F_7Fe(CO)_4I + CO$$

$$[(C_6H_5)_3P]_3RhCl + C_2H_4 \longrightarrow (\pi\text{-}C_2H_4)Rh[P(C_6H_5)_3]_2Cl + (C_6H_5)_3P$$

$$[(C_6H_5)_3P]_2Ir(CO)Cl + CH_3I \longrightarrow CH_3Ir[P(C_6H_5)_3]_2(CO)(Cl)(I)$$

$$[(C_6H_5)_3P]_3Pt + CH_3C{\equiv}CCH_3 \longrightarrow [\pi\text{-}(CH_3)_2C_2]Pt[P(C_6H_5)_3]_2 + (C_6H_5)_3P.$$

Principal reactions. Each class of organometallic complex (see Table 58) has its own characteristic chemical and physical properties associated with the presence of a particular type of organic group. The degree to which group properties are shown, however, depends on the firmness with which the ligand is attached to the metal atom. Consequently, there are no reactions generally applicable to all those compounds, and even for a particular class large gradations in reactivity are observed.

Many organometallics are endothermic compounds; that is, their formation from the constituent elements involves an increase in free energy. Hence there is an inherent thermal instability with respect to decomposition into the metals and the hydrocarbons that make them up. Nevertheless, many hundreds of these compounds can be prepared, and in some cases they show considerable resistance to thermal decomposition, because it is necessary for a favourable free-energy change to be accompanied by a reaction path of sufficiently low energy if decomposition is to occur at a measurable rate. In many cases, decomposition may be kinetically rather than thermodynamically controlled—that is, it may not be observed normally because the rate is too slow, even though the energy requirements are favourable. The limiting process may be the initial breaking of an M—C bond to form M· and R· radicals (homolytic dissociation) or M^+ and R^- ions (heterolytic dissociation). Such bond dissociations are likely to be facilitated when empty low-lying orbitals are present in the metal atom, since electrons can temporarily reside in these. In both homolytic and heterolytic processes, the carbon fragments are generally reactive and rapidly form stable products; for example, by dimerization. It is the formation of these more stable products that makes the decomposition irreversible; but if there is a high activation energy (energy needed to initiate the process) for the dissociation, decomposition will not occur.

All organometallic compounds are thermodynamically unstable to oxidation because of the large favourable free-energy change in the conversion into metal oxide, carbon dioxide, and water. In this respect organometallic compounds are no different from purely organic compounds except that with the former an additional product (metal oxide) is involved. Again, however, kinetic factors (considerations of reaction rate) are often decisive both in enhancing stability toward oxygen and in lowering resistance to oxidative attack. In this case also instability can be associated generally with the presence of empty low-lying orbitals. Similar considerations apply to hydrolytic stability. Hydrolysis often involves an initial bonding with a water molecule, a process facilitated by an empty metal orbital. The M—C bond polarity, however, also is an important factor in hydrolysis because a further step in the process is the elimination of R^- as a hydrocarbon RH.

The existence of a closed-shell (18-electron) configuration has a bearing on the reactivity of a complex (see above). For complexes in which the metal atom already follows the effective-atomic-number rule, reaction may require displacement of one or more ligand molecules in order to provide a vacant coordination site or sites for a reactant molecule. This situation may lead only to the exchange of one ligand for another, but often a molecular rearrangement follows, affording an entirely new structure.

Cleavage reactions

Cleavage and insertion reactions are the hallmark of σ-bonds between transition metals and alkyl or aryl groups. Cleavage reactions are those in which the C—M bond is split, with attendant formation of a new bond to an inorganic atom. Typical reagents that may effect cleavage are acids, hydrogen, halogens, and metal halides.

$$(CH_3)_2Pt[P(C_2H_5)_3]_2 + HCl \longrightarrow CH_3Pt[P(C_2H_5)_3]_2Cl + CH_3Cl$$

$$C_6H_5Pt[P(C_2H_5)_3]_2Cl + H_2 \longrightarrow HPt[P(C_2H_5)_3]_2Cl + C_6H_6$$

$$(C_6H_5)_2Ti(\pi\text{-}C_5H_5)_2 + 2Cl_2 \longrightarrow Cl_2Ti(\pi\text{-}C_5H_5)_2 + 2C_6H_5Cl$$

$$(C_6H_5)_3Cr(THF)_3 + 3HgCl_2 \longrightarrow 3C_6H_5HgCl + CrCl_3 + 3THF$$

The term insertion designates a reaction in which a C—M bond is broken, with the reactant molecule's being placed between the two parts and joined to both. This process is the basis of the Ziegler–Natta olefin polymerizations (see above), but it also is known in many simpler forms.

$$C_6H_5Fe(CO)_2(\pi\text{-}C_5H_5) + (C_6H_5)_3P \longrightarrow C_6H_5COFe(CO)[P(C_6H_5)_3](\pi\text{-}C_5H_5)$$

$$CH_3Mn(CO)_5 + SO_2 \longrightarrow CH_3SO_2Mn(CO)_5$$

Conversion of σ- to π-complexes

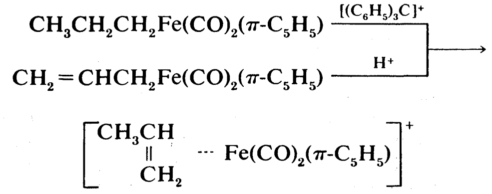
In many reactions, a σ-bonded organic group is converted into a π-complexed group. A simple example involves the complex $\sigma\text{-}CH_2{=}CHCH_2Mn(CO)_5$, which on heating or exposure to ultraviolet light releases carbon monoxide and forms $(\pi\text{-}C_3H_5)Mn(CO)_4$. Similar decarbonylations (loss of carbonyl groups) have been observed with σ-allyl carbonyl derivatives of cobalt, iron, chromium, molybdenum, and tungsten, in each case π-allyl complexes of these metals being formed.

In 1919 and afterward many polyphenylchromium compounds were prepared by treating chromium chloride ($CrCl_3$) with phenylmagnesium bromide. In 1954 it was shown that several of these compounds were cationic arene–chromium π-complexes; *e.g.*,

$$[(\pi\text{-}C_6H_6)Cr(\pi\text{-}C_6H_5C_6H_5)]^+,$$

which had been formed by rearrangement of $\sigma\text{-}C_6H_5Cr$ groups either in a hydrolysis step or thermally. Most of these transformations are exceedingly complex, and a multitude of products is formed. The history of the polyphenylchromium complexes is an example of the failure to recognize the true nature of a series of compounds because of the lack of sufficiently sophisticated analytical techniques.

Several σ-alkyl complexes have been converted to cationic olefin π-complexes by hydride abstraction (removal of H^-) or by protonation (addition of H^+).



$$CH_3CH_2CH_2Fe(CO)_2(\pi\text{-}C_5H_5) \xrightarrow{[(C_6H_5)_3C]^+}$$
$$CH_2{=}CHCH_2Fe(CO)_2(\pi\text{-}C_5H_5) \xrightarrow{H^+}$$
$$\longrightarrow \left[\begin{matrix} CH_3CH \\ \| \\ CH_2 \end{matrix} \cdots Fe(CO)_2(\pi\text{-}C_5H_5) \right]^+$$

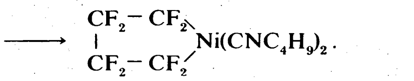
An unusual group of compounds contains a fluorocarbon group (chemical unit of only carbon and fluorine atoms) σ-bonded to a transition metal. These complexes are more stable chemically than their alkyl or aryl counterparts—*e.g.*, $C_2F_5Mn(CO)_5$ decomposes at 150° C (302° F), whereas $C_2H_5Mn(CO)_5$ decomposes at 25° (77° F); $C_6F_5Mn(CO)_5$ decomposes at 170° C (338° F), whereas $C_6H_5Mn(CO)_5$ decomposes at 100° C (212° F). Thermodynamic and kinetic factors both are thought to be responsible for the enhanced stability of the fluorocarbon compounds. The fluorocarbon–metal bonds are more ionic than hydrocarbon–metal bonds; that is, they show more complete association of the shared electron pair with the carbon atom of the bond. This effect is due to the strongly electron-withdrawing character of the fluorocarbon groups, which tend to pull the electron pair farther from the metal atom. Moreover, alkyl–metal bonds frequently decompose by a reaction that is the reverse of an insertion reaction, $M{-}C_2H_5 \rightarrow M{-}H + C_2H_4$. The analogous reaction involving elimination of a fluoroolefin and formation of a metal fluoride, although thermodynamically favoured, requires more activation energy in order to proceed at a measurable rate. Because of their stability, bonds between fluorocarbon groups and transition metals show little or no tendency to undergo insertion reactions. More than 500 complexes of this type are known, produced by reactions such as

$$HMn(CO)_5 + CF_2{=}CF_2 \longrightarrow HCF_2CF_2Mn(CO)_5$$

$$NaRe(CO)_5 + C_6F_6 \longrightarrow C_6F_5Re(CO)_5 + NaF$$

$$(\pi\text{-}C_5H_5)_2ZrCl_2 + 2C_6F_5Li \longrightarrow (C_6F_5)_2Zr(\pi\text{-}C_5H_5)_2 + 2LiCl$$

$$(C_4H_9NC)_4Ni + 2CF_2{=}CF_2$$



$$\longrightarrow \begin{matrix} CF_2-CF_2 \\ | \\ CF_2-CF_2 \end{matrix} \!\!\!>\! Ni(CNC_4H_9)_2 .$$

Reactions of π-complexes

The reactions of π-complexes are extremely extensive; it has been said that the cyclopentadienyl group forms more complexes than any other organic ligand. Whether or not this is strictly true, the range of complexes formed by this five-membered ring is certainly great. Simple cyclopentadienyl derivatives, $(\pi\text{-}C_5H_5)_2M$, of the first row transition elements, from vanadium through nickel, all have the same structure, consisting of a metal atom sandwiched between two five-membered rings. This situation gives rise to a fairly constant melting point (173° C [343° F]), although the differing electronic configurations result in markedly different reactivities. Thus, in marked contrast with the highly stable ferrocene, chromocene bursts into flame spontaneously in air.

One of the main factors responsible for the many ramifications of coordination chemistry is the ability of a metal atom to bond more than one type of ligand within its coordination sphere at the same time. Moreover, certain ligands stabilize the bonding of others, and this is especially true of the $\pi\text{-}C_5H_5$ group, as illustrated by the existence of the following compounds of metals of the first transition series:

$$(\pi\text{-}C_5H_5)_2Ti(C_6F_5)_2,$$

$$(\pi\text{-}C_5H_5)V(\pi\text{-}C_7H_7),$$

$$(\pi\text{-}C_5H_5)Cr(NO)_2Cl,$$

$$(\pi\text{-}C_5H_5)Mn(\pi\text{-}C_6H_6),$$

$$(\pi\text{-}C_5H_5)Fe(\pi\text{-}C_4H_4N),$$

$$(\pi\text{-}C_5H_5)Co(\pi\text{-}C_5H_6),$$

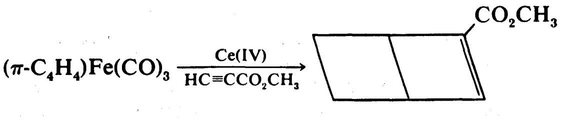
$$(\pi\text{-}C_5H_5)Ni(CO)CF=CFCF=CF_2.$$

The cyclopentadienylmetal carbonyls undergo a number of chemical reactions and show several resemblances to the metal carbonyls. For example, the $(\pi\text{-}C_5H_5)M$ group is isoelectronic with (*i.e.,* has the same number of electrons as) the $M'(CO)_3$ group, where M′ is the element preceding M in the periodic table. Many complexes containing the $Fe(CO)_3$ group thus have counterparts containing the $(\pi\text{-}C_5H_5)Co$ group; similarly, the chemistry of $[(\pi\text{-}C_5H_5)Fe(CO)_2]_2$ bears many resemblances to that of $[Mn(CO)_5]_2$, and that of $(\pi\text{-}C_5H_5)Mn(CO)_3$ to $Cr(CO)_6$. The cyclopentadienylmetal carbonyls also resemble the metal carbonyls in that in both cases there are numerous polynuclear complexes containing metal–metal bonds; *e.g.,* $[(\pi\text{-}C_5H_5)Ni]_3(CO)_2$, $[(\pi\text{-}C_5H_5)RhCO]_3$, or $[(\pi\text{-}C_5H_5)FeCO]_4$.

Aromatic reactions of π-complexes

A striking feature of the chemistry of certain cyclopentadienyl complexes is the ability of the five-membered ring to undergo reactions typical of aromatic systems. Such chemisty was first demonstrated for ferrocene, which can be acylated (substituted with an acyl group) under Friedel-Crafts conditions—that is, with aluminum choloride as catalyst. Furthermore, one of the hydrogen atoms can be replaced by a lithium atom or a mercuriacetate group, using butyllithium or mercuric acetate, respectively. Introduction of these and other functional groups has allowed the synthesis of hundreds of ring-substituted ferrocene compounds. Many other cyclopentadienylmetal compounds, on the other hand, do not undergo substitution reactions for either of several reasons: (1) the metal atom interferes with the reaction—*e.g.,* becomes oxidized; or (2) the complex does not survive the reaction conditions. Aromatic behaviour also has been shown by $(\pi\text{-}C_5H_5)_2M$, M=Ru, Os; $(\pi\text{-}C_5H_5)Mn(CO)_3$, termed "cymantrene"; $(\pi\text{-}C_5H_5)Cr(CO)_2NO$; $(\pi\text{-}C_5H_5)_2ReH$; and certain other compounds.

In tricarbonyl(cyclobutadiene)iron, the metal–ligand bonding is sufficiently strong to allow the cyclobutadiene group to show aromaticity. For example, CH_3CO and HgCl groups may be substituted for H atoms of the ring to give useful intermediates for further syntheses. On the other hand, $(\pi\text{-}C_4H_4)Fe(CO)_3$ can be oxidized, releasing cyclobutadiene, which may be characterized by the reactions it undergoes immediately.



When this reaction is carried out with $C_6H_5C{\equiv}CH$, the organic compound produced is in the unusual situation of having a classical benzene ring bonded to its valence isomer, Dewar benzene (named for the Scottish chemist who proposed this structure for ordinary benzene), the structure shown in the above formula.

Analytical chemistry. Analytical methods for the organic derivatives of the transition metals are generally similar to those for the organic derivatives of the main-group metals. Spectroscopic techniques are an important source of information, and with metal carbonyl derivatives, infrared spectroscopy is extremely valuable in determining the number of carbonyl groups and their arrangement around the metal. Proton magnetic resonance spectra are of help in interpreting the nature of the bonded hydrocarbon groups, although the number of fluxional systems discovered means that care must be exercised in the interpretation of apparently simple spectra. Other spectroscopic methods—*e.g.,* Mössbauer, and electronspin resonance techniques—have been used where appropriate.

The increasing complexity of the compounds discovered makes X-ray crystallographic methods the final arbiter of structures; hundreds of new structures are determined by this method each year. This analytical technique has been greatly facilitated by computer processing of data.

Industrial applications. Relatively few organometallic compounds of the transition metals find direct practical application. On the other hand, their value as reaction intermediates is almost inestimable.

Use in the petroleum industry

As already indicated, organometallics are widely employed in the petroleum-based industries. Certain industrial-scale syntheses depend upon the transient existence of catalysts in the form of metallo-organic complexes, which are constantly regenerated. Hydroformylation, or the conversion of olefins into aldehydes or ketones using carbon monoxide and hydrogen under pressure (100°–300° C [212°–572°F]; 100 atm), is catalyzed by cobalt carbonyl and related complexes, including those of rhodium. The catalytic material in the system is thought to be a hydridocarbonyl—such as $HCo(CO)_3$—derived from $Co_2(CO)_8$, and the various reaction steps include insertion reactions (olefin inserted into metal–hydrogen bonds and CO into metal–alkyl bonds) as well as oxidative-addition and reductive-elimination reactions. The overall process clearly depends upon organometallic intermediates and their cleavage reactions.

$$Co_2(CO)_8 + H_2 \rightleftharpoons 2HCo(CO)_3 + 2CO$$

$$\xrightarrow[\pi\text{-complex formation}]{RCH=CH_2} (\pi\text{-}RCH=CH_2)Co(CO)_3H$$

$$\xrightarrow{\text{insertion}} RCH_2CH_2Co(CO)_3$$

$$\xrightarrow[\text{insertion}]{CO} RCH_2CH_2COCo(CO)_3 \xrightarrow[\text{oxidative addition}]{H_2} RCH_2CH_2COCo(CO)_3H_2$$

$$\xrightarrow[\text{elimination}]{\text{reductive}} \underset{\text{product}}{RCH_2CH_2CHO} + \underset{\textit{catalyst}}{HCo(CO)_3}$$

The mechanisms of such reactions are related to the mechanisms of catalytic hydrogenation of olefins or acetylenes by certain metal complexes. In both cases, the catalyst carries out three basic reactions: (1) activation of molecular hydrogen by breaking of the H—H bond and generation of two M—H bonds; (2) activation of the olefin by formation of a π-complex; and (3) transfer of hydrogen to the olefin (by insertion to form a σ-metal–alkyl group, followed by cleavage of this group by hydrogen). It is apparent that conversion of a π-complex into a σ-complex represents a key step in the overall process.

Use in hydrolysis reactions

Another useful synthesis depends in part for its success on the fact that coordinated olefins become susceptible to a hydrolysis reaction that leads to oxidation. In this way ethylene is converted to acetaldehyde; propylene, to acetone; and ethylene and acetic acid to vinyl acetate (which serves as a monomer for polymerization). In one variation of this basic reaction, ethylene and oxygen are passed together through an aqueous solution of palladium and cupric salts. The overall reaction ($C_2H_4 + {}^1/_2O_2 \rightarrow CH_3CHO$) results from a combination of three steps:

$$(1)\quad C_2H_4 + PdCl_2 + H_2O \longrightarrow CH_3CHO + Pd + 2HCl$$

$$(2)\quad Pd + 2CuCl_2 \longrightarrow PdCl_2 + 2CuCl$$

$$(3)\quad 2CuCl + 2HCl + \tfrac{1}{2}O_2 \longrightarrow 2CuCl_2 + H_2O.$$

Reaction (1), the reaction that forms the product, is itself composed of several steps, including formation of the

π-complex anion $[(\pi\text{-}C_2H_4)PdCl_3]^-$ (related to the anion of Zeise's salt) and its rearrangement after hydrolysis to a σ-complex $[CH_2OHCH_2PdCl_3]^{2-}$.

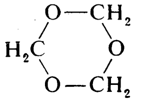
Other industrial reactions catalyzed by transition-metal complexes include the isomerization, dimerization, and polymerization of olefins, the trimerization of acetylenes, the tetramerization of acetylene (to cyclooctatetraene), and the dismutation (conversion to products with both larger and smaller carbon skeletons) of olefins (*e.g.,* the conversion of propene into ethylene and 2-butene). In each of these reactions the making and breaking of carbon–metal bonds are essential steps. (M.I.B./F.G.A.S.)

POLYMERS

Polymers are essential to the existence of man and his civilization. The human body, all animal and plant tissues, and most structural substances of organic nature—such as proteins, wood, resin, and chitin, the hard coating of insects—consist of polymeric (macromolecular) materials. Many minerals, such as diamond, quartz, and feldspar, are inorganic polymers, and numerous products of ancient and modern industry, such as concrete, porcelain, glass, textiles, paper, rubbers, and plastics, are either entirely or substantially polymeric. Only in the first half of the 20th century, however, was it clearly recognized that these substances all possess one essential common feature, namely, that they consist of very large (macro) molecules.

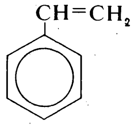
Definition

A polymer (Greek *polys,* "many"; *meros,* "part" or "unit") is a substance the molecules of which are multiples of low-molecular-weight units, molecular weight being, as the name suggests, a measure of the weight of a molecule relative to a chosen standard (for practical purposes, the weight of an atom of hydrogen). The low-molecular-weight unit making up the polymer is referred to as the monomer. As long as a polymer is strictly uniform in molecular weight and molecular structure, its degree of polymerization is indicated by the Greek word for the number of monomers it contains; thus, for example, dimers, trimers, tetramers, and pentamers indicate twofold, threefold, fourfold, and fivefold polymers. The term polymer, however, is used to designate a combination of an unspecified number of monomer units. Trioxymethylene, for example, is a (cyclic) trimer of formaldehyde (CH_2O):

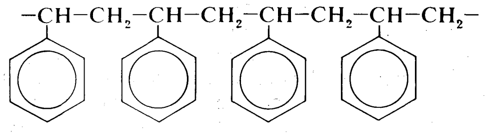


trioxymethylene

and polystyrene is the polymer prepared from the monomer styrene:



styrene



polystyrene

If the number of monomer units in a polymer becomes very large, the latter is sometimes called a high polymer. According to present-day usage, a polymer or high polymer need not consist of individual molecules all having exactly the same molecular weight or precisely the same chemical composition and molecular structure.

In the case of some natural polymers, such as the proteins, the individual molecules all have the same molecular weight and molecular structure. This identity of the polymer molecules results from highly specific cellular processes that produce them according to fixed patterns. With most synthetic and natural polymers, however, significant differences occur in the molecular weight of the individual macromolecules, with the result that the material itself is a mixture of homologous polymeric constituents—that is, of individual molecules differing in the number of monomer units they contain. Methods are not available for preparing synthetic polymers of fixed molecular weight, nor can the prepared mixture be separated completely into fractions with uniform molecular weights. Variability in the composition and molecular structures of polymers results from the presence of different end groups that terminate the chains, occasional branches in the chain, variations in the arrangement of the monomeric units, and irregularity in the sequence of monomeric units of different types if more than one type is present.

A completely rigorous system of classifying and naming polymers, conforming to the practice in naming ordinary organic and inorganic molecules, is impossible because of the common variations in structure between individual molecules of most polymer types—variations that cannot be eliminated, or even determined quantitatively in nature or amount by available methods. Nevertheless, it is convenient to treat a particular type of polymer as a single molecular species for purposes of discussion because these variations in structure do not significantly affect the physical and chemical properties of the substance and do not affect its practical application.

NATURAL POLYMERS

All living beings—animals and plants alike—are composed to a very large extent of organic polymers, which not only provide structural materials for the maintenance of life functions but also participate in carrying out the life functions themselves. The solid portions of all plants consist of several polymers, the most important of which are cellulose, lignin, and various resins.

Cellulose and lignin

Cellulose in wood is intimately connected with lignin; but in other parts of plants, it is found in relatively pure form as fibres such as cotton and flax. Cellulose itself is a polysaccharide, that is to say, a polymer whose monomer units are individual sugar molecules. In cellulose the sugar glucose comprises the monomer units, and they are joined to give molecular weights of up to several millions. In most fibres the polymer chains are combined to give laterally ordered crystalline units that, in turn, are separated by disordered or amorphous areas. This two-phase structure gives cellulose fibres their strength and, at the same time, their softness and pliability.

In all types of wood cellulose is accompanied by lignin, a second polymer, which shows much less regularity than cellulose. Lignin molecules consist of a complicated three-dimensional network of phenolic residues joined chemically by means of short hydrocarbon chains. Because of this molecular complexity, the molecular weight of lignin is not exactly known, but in any event it is very high. Wood also contains several types of resins, which are polymers of the simple hydrocarbon isoprene and have molecular weights ranging from a few thousand to several hundred thousand.

The most important terpene, as the isoprene polymers are called, is rubber, present in the sap of many tropical trees, especially of the genus *Hevea.* Rubber is a polyhydrocarbon and consists, in the native state, of long chains of isoprene, which have molecular weights around 500,000. Balata, another natural, rubberlike resin, also is a polyisoprene; it differs from hevea rubber chiefly in the geometry of double bonds, or linkages, between carbon atoms in the molecular chains.

Life on Earth is based largely on the existence, properties, and reactions of two classes of polymers—the proteins and the nucleic acids. The monomeric units of proteins

Proteins and nucleic acids

are amino acids, relatively small molecules containing amino ($-NH_2$) and carboxylic acid ($-COOH$) groups. The twenty amino acids that occur normally in proteins are always arranged in precise, but different, orders in different protein molecules. Certain proteins have molecular weights around 50,000; in others the molecular weight reaches several millions. All fibres and membranes of animal tissues consist of high-molecular-weight proteins; the processes of life—such as metabolism, vision, and respiration—are based on the protein materials known as enzymes, which act as catalysts for biochemical reactions. Generally, the exterior parts of the animal body—fur, hair, nails, horn, and skin—consist of proteins with molecular structures of three-dimensional networks of very high molecular weight (*e.g.,* keratin).

Other important natural polymers are the starches, which—like cellulose—are polymers of glucose. Starches are widely distributed in plants and play an important role as a food for animals of all kind. Starches differ from cellulose chiefly in the way the individual monomeric units are joined in the molecules; some varieties—amylose, for example—are simple chains with molecular weights around 100,000; others—called amylopectins—are three-dimensional networks (like lignin) with very high molecular weights. All hard, tough, and strong parts of insects and crustacea—shells, wings, and claws, for example—consist of a nitrogen-containing polysaccharide called chitin. Chitin is composed of long chains of glucose molecules that have been modified by acetylated amino groups; the molecular weights of typical chitin polymer chains are around 100,000.

Diamond, graphite, and silica

Many inorganic polymers also are found in nature; these are characterized by their hardness, high melting range, and great mechanical strength. Some inorganic polymers are formed by atoms of the element alone; in others the monomeric units are inorganic compounds. Diamonds are a good example of the first class; this substance consists of a three-dimensional network of carbon atoms that are combined with each other exclusively by single bonds. In fact, diamond is the simplest crystalline, three-dimensional polymer that exists. Another polymeric form of the element carbon is graphite, the writing material in a lead pencil. Graphite consists of two-dimensional polymeric layers of carbon atoms joined together alternately by single and double bonds. Between these layers of densely packed carbon atoms are relatively large distances, a factor that causes the typical lamellar (layer-like) character of all graphitic materials. Elemental sulfur and selenium exist generally in crystalline modifications but also in a polymeric form consisting of long chains with molecular weights around 100,000.

The most important inorganic polymer is silica, or silicon dioxide, which exists in several crystalline modifications and in a glossy amorphous form; they are all three-dimensional networks—partly regular, partly random—held together by chemical bonds between silicon and oxygen. The strength of the silicon–oxygen bonds is responsible for the high softening range and the extreme hardness of these compounds.

The three-dimensional silica networks are capable of accepting electrons; in the process they are converted into polymeric ions that can be neutralized and stabilized by positively charged metal ions, such as those of sodium, potassium, magnesium, calcium, and aluminum. The result is a large and important family of silicates that, together with other metals, form most of the hard and high-melting natural rocks—including granite, feldspar, and agate.

SYNTHETIC POLYMERS

Organic. In the course of their studies, organic chemists of the 19th century frequently encountered materials that were neither ordinary liquids nor crystalline solids but of an amorphous semisolid or glassy character. These materials generally defied the classical methods of purification, isolation, and identification. As a consequence, most of these substances were only referred to in passing in scientific writings but were not further investigated. In a few instances, however, the yield of such a material was so considerable and its properties were so attractive for practical use that its preparation was more thoroughly studied, and a larger scale development of the product was carried out. Of these developments, probably the most important stemmed from the discovery in 1909 by Leo Baekeland, a Belgian-born U.S. chemist, of a viscous product from the reaction of phenol with formaldehyde. Baekeland discovered that many species of useful plastics could be prepared by varying the proportions of the components and the conditions of the reaction. This discovery led to the development of the Bakelite resins, which occupy an important position in the plastics industry. In the 1920s, when the structure and character of the most important natural polymers—cellulose, proteins, and rubber—were elucidated, a new, broadly based series of systematic studies of polymer formation was undertaken. These studies led to the production of a large number of synthetic polymers, which in turn permitted an amazing variety of industrial applications of polymers as fibres, plastics, synthetic rubbers, coatings, and adhesives.

This rapid development in the chemistry of synthetic polymers was aided by the fact that most building units for the preparation of synthetic polymers are products of the petroleum-refining industry, which also was undergoing rapid development at the same time.

It is convenient to classify the synthetic organic polymers according to the chemical character of their monomers, beginning with the simplest variety, the hydrocarbons.

Polyhydrocarbons. Many simple and inexpensive unsaturated hydrocarbons, such as ethylene, propylene, butadiene, and styrene, can be readily transformed into polymeric materials by simple addition polymerization, the repeated addition of one monomer unit after another to the growing polymer chain. In this way important thermoplastic and elastomeric (polymers with elastic properties) products are obtained.

Polyethylenes and polypropylenes

There are two ways to prepare polyethylene on an industrial scale: (1) polymerization with oxygen as catalyst at high pressure (1,000–2,000 atmospheres) and high temperatures (190°–210° C [374°–410° F]) and (2) polymerization with organometallic catalysts at ordinary pressure and ordinary temperatures. The high-pressure polyethylenes have specific gravities between 0.92 and 0.93 (water = 1.000), melting points around 110° C (230° F), and molecular weights from 20,000 to 40,000. They represent relatively soft, but tough, thermoplastic materials with excellent electrical and mechanical properties and with outstanding resistance against most acids, bases, and solvents. Such high-pressure polyethylenes are widely used for the production of synthetic fibres and for the preparation of bottles and small containers. The low-pressure polyethylenes have specific gravities between 0.95 and 0.97, melting points around 134° C (273° F), and molecular weights up to many hundred thousands. They represent hard, tough, strong thermoplastic materials with excellent electrical, mechanical, and chemical properties. They are widely used for the production of larger bottles, containers, pails, and trash cans.

There are two varieties of polypropylene: a soft, rubbery type of minor importance and a hard, crystalline, high-melting type, which is a valuable thermoplastic material of wide applications in certain areas of the textile industry and also used to make molded objects. The second type of polypropylene has crystalline character, melts at around 178° C (352° F), and has a density of about 0.90, and a molecular weight ranging from 50,000 to 200,000.

Polybutadines and polystyrenes

Most polybutadienes are typical elastomers; they have many valuable properties. Butadiene can be polymerized in solution, suspension, or emulsion with different types of catalysts. Some of the products are crystalline, others are amorphous. All of them remain soft and resilient down to low temperatures and can be cured into elastomers of great strength and excellent abrasion resistance. The molecular weights range from 80,000 to 200,000. Most polybutadienes are soluble in aromatic solvents, such as benzene.

Styrene polymerizes readily in pure form or in solution, suspension, or emulsion—usually under the influence of peroxide catalysts. The polymer is a colourless, transparent, thermoplastic resin (polymer product used to make objects of plastic with the application of heat), which has a specific gravity around 1.08 and a molecular weight range from 50,000 to 200,000. Commercially produced polystyrene is amorphous, has useful electrical and mechanical properties, and softens around 90° C (194° F). It is highly resistant to moisture, acids, and bases but is easily attacked by organic solvents. Polystyrene and closely related polymers are the most popular resins for certain molding and extrusion procedures in the forming of plastic objects.

Particularly important are copolymers (polymers composed of several monomer units) of butadiene and styrene, which are made in emulsion with the aid of peroxide catalysts. During the 1940s and 1950s these amorphous elastomeric materials were the most important synthetic rubbers and, subsequent to that time, continued to be valuable and useful products. They can be compounded and used like natural rubber, and they exhibit excellent resistance to abrasion and deterioration caused by oxygen, heat, and light.

Polyvinyls and polyacrylics. Many derivatives of ethylene still retain the double bond and are able to polymerize. Among the materials produced in this way are several important families of polymers, the polyvinyls and polyacrylics.

Polyvinyl
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The replacement of one hydrogen atom in ethylene by chlorine leads to vinyl chloride, which polymerizes readily under the influence of peroxide catalysts in suspension or emulsion. The product, polyvinyl chloride, is a colourless, hard, tough thermoplastic material, which softens only at a high temperature. Its specific gravity is around 1.55; it is substantially amorphous and resistant to most destructive chemical agents, including hot organic solvents. The molecular weight ranges from 50,000 to 150,000. It is a very useful material prepared as sheets and as extruded or molded products.

Vinyl acetate and vinyl butyrate are liquid esters of vinyl alcohol (the hydroxide equivalent of vinyl chloride). They polymerize with peroxide catalysts in solution, suspension, and emulsion to give amorphous, soft resins with molecular weights between 30,000 and 100,000. The products find many uses as coatings and adhesives. Both vinyl acetate and vinyl butyrate copolymerize readily with vinyl chloride to give a large family of thermoplastic materials. In these substances the vinyl chloride contributes hardness and high-melting characteristics, whereas the vinyl esters act as internal (*i.e.,* chemically bonded) plasticizers, additives that impart favourable qualities to plastics. These copolymers exhibit a wide range of properties; hence they find many applications, such as for coatings, cohesives, films, and extruded or molded articles of many kinds.

oly-
crylics

The basic units for the large and important family of polymeric materials known as polyacrylics are acrylic acid (propenoic acid) and methacrylic acid (a methylacrylic acid with the methyl and carboxyl groups joined to the same carbon atom). The methyl esters of these acids are readily polymerized with the aid of peroxide catalysts in the pure material or in solution, suspension, or emulsion. Polymethyl methacrylate is a hard, colourless, brilliantly transparent thermoplastic material that is widely used to manufacture extruded and molded articles. Equally important is the polymer of acrylonitrile (vinyl cyanide) that is usually made in suspension or emulsion with the aid of peroxide catalysts. The product has a molecular weight in the range 70,000 to 120,000; it represents a white, hard, moderately crystalline, material, which softens at a high temperature and is soluble only in a few selected high-boiling organic liquids, such as *N,N*-dimethylformamide and dimethyl sulfoxide. Copolymers of acrylonitrile (80–90 percent) and methyl acrylate or vinyl acetate occupy an important place in the man-made fibre industry. Copolymers of acrylonitrile (10–15 percent) and styrene represent thermoplastics for extrusion and molding. The products possess superior hardness, higher softening ranges, and better impact resistance than the polymers prepared from either monomer alone. Finally, copolymers of acrylonitrile (20–30 percent) and butadiene, prepared in emulsion, represent a valuable species of synthetic elastomer. These so-called nitrile rubbers have excellent resistance to hot oils, oxygen, ozone, light, and heat; therefore, they find important applications in the construction of airplanes, automobiles, and electrical equipment.

Polyacetals and polyethers. In addition to the polymers whose backbone chains consist of a continuous sequence of carbon–carbon bonds, there are many important polymers with oxygen or nitrogen atoms in the backbone chain along with carbon atoms. Among those with oxygen atoms in the backbone are the polyacetals and polyethers. One of the simplest polyacetals is polyformaldehyde, a hard, high-melting, crystalline material with excellent resistance against abrasion and against the action of solvents. It is prepared by polymerization of formaldehyde—a very simple and inexpensive monomer—in the gaseous state with anionic catalysts. It is used for making machine parts in place of, or in conjunction with, certain metals. Polyoxyethylene, a polyether, is prepared in bulk or in solution from ethylene oxide with cationic catalysts. The product is a white, crystalline, water-soluble material with a melting point of about 75° C (167° F); it is a particularly useful thickener of aqueous solutions and a material used for coatings for textiles and paper.

Polyesters. When a bifunctional carboxylic acid, such as adipic acid, $HO_2C(CH_2)_4CO_2H$, reacts with a bifunctional alcohol, such as ethylene glycol, $HO(CH_2)_2OH$, under conditions that lead to the elimination of water, a linear macromolecule is formed that is characterized by a repetition of ester groups ($-COO-$) along the chain and is called, therefore, a polyester. Each combination of a specific dicarboxylic acid and a specific glycol leads to a particular polyester.

Nomen-
clature
of linear
polymers

In a convenient naming system for linear polyesters, the number of carbon atoms in the glycol and in the dicarboxylic acid are used after the word polyester. For example, hexamethylene glycol, $HO(CH_2)_6OH$, with six carbon atoms, and adipic acid, $HOOC(CH_2)_4COOH$, also with six carbon atoms, give a polyhexamethylene glycol adipate commonly referred to as polyester-6,6. Similarly, polydecamethylene succinate is polyester-10,4. Polyesters can also be formed by the self-condensation of hydroxycarboxylic acids, such as hydroxyundecanoic acid, $HO(CH_2)_{10}COOH$. The resulting polymers need only one figure to characterize them; in this case, the product polymer is called polyester-11.

All linear polyesters made with bifunctional open-chain glycols and dicarboxylic acids are colourless, crystalline, thermoplastic materials. Their melting points and solubilities vary over a wide range and depend on the length of the methylenic ($-CH_2-$) segments between the ester bonds in the chain. Because of the polarity (separation of electrical charge) of these bonds, they are mainly responsible for the intermolecular attraction between the chains, and hence for their softening and solubility behaviour. The closer the ester bonds, the higher is the melting point of the polyester and the less is its solubility. Polyethylene glycol oxalate, polyester-2,2, melts around 200° C (392° F), whereas polyester-6,6 melts around 80° C (176° F). The former is soluble in only a few selected organic liquids; the latter dissolves readily in most ordinary organic solvents, such as benzene, ethyl methyl ketone, and butyl acetate.

Low-molecular-weight polyesters (molecular weight up to 5,000) are excellent plasticizers for many natural and synthetic resins; the high-molecular-weight species (above 10,000–15,000) are valuable thermoplastics themselves and are used for the manufacture of films and molded objects. Polyesters with high melting points (250° C [482° F] and above) are obtained from aromatic (cyclic) constituents. Examples are the polymer made from ethylene glycol and terephthalic acid (benzene-1,4-dicarboxylic acid)—the so-called polyester-2T—and the material made by the self-esterification of 4-hydroxybenzoic acid—the HOB-polyester—both of which melt above 250° C and do not dissolve in most organic solvents. Because of their high melting points and insolubility, they represent valuable materials for the formation of films used in electrical insulation, photography, and packaging, and also for fibres used in fabrics and tire cord. When the polyesterification is carried out in the presence of a polyfunctional component—such as pentaerythritol, a tetrahydroxy compound, or a tricarboxylic acid—the product is not a linear macromolecule but a highly branched species that, after sufficient removal of water, is converted into a three-dimensional network. Unlike the thermoplastic linear polyesters, systems of this kind are thermosetting—that is, the original material softens on heating, but once set in a new shape it is extremely resistant to further heating. The thermoset polyesters are insoluble, hard, and tough materials of extremely high molecular weight. They are widely used as coatings, laminating agents, and in making molded objects. They are usually applied as thermosetting compositions that are cured into the thermoset state.

Nylons

Polyamides. When a bifunctional amine (two $-NH_2$ groups) reacts with a bifunctional acid (two $-CO_2H$ groups) with elimination of water, a linear macromolecule is formed that contains in the chain a regular repetition of amide groups ($-CO-NH-$) and is therefore called a polyamide. More commonly, these materials are known as nylons. The same nomenclature is used for polyester as for polyamides. Thus, the nylon made of hexamethylenediamine, $H_2N(CH_2)_6NH_2$, and adipic acid, $HOOC(CH_2)_4COOH$, is nylon-6,6; the polymer made from the same diamine and sebacic acid, $HOOC(CH_2)_8COOH$, is nylon-6,10; and the material made by the self-polymerization of aminocaproic acid, $H_2N(CH_2)_5COOH$, is nylon-6. All polyamides consisting of linear, bifunctional components are hard, tough, colourless, crystalline, thermoplastic materials. Their softening and solubility behaviour is controlled by the magnitude of the intermolecular forces between the individual chains; the major contribution to these forces comes from the amide bonds ($-CO-NH-$), which are polar themselves and also capable of establishing hydrogen bridges ($\gt C{=}O \cdots H{-}N\lt$) between adjacent macromolecular chains. The closer the amide groups are in the chain, the stronger is the cohesion between chains, and the stronger the cohesion, the higher the melting point, and the greater the insolubility. For example, the polyamide of ethylenediamine and oxalic acid, nylon-2,2, melts above 300° C, or 572° F (with decomposition), whereas nylon-6,6 melts around 260° C (500° F) and nylon-6,10 at 215° C (419° F). An interesting contrast exists between the polyester series and the polyamide series: corresponding members—such as the polyester-6,6 and the polyamide-6,6—differ substantially in softening temperature and solubility, the nylons having much higher melting points and much lower solubilities than the esters. This difference is usually attributed to the fact that the polyamides can develop hydrogen bridges whereas the polyesters cannot.

Although nylons find many important applications in the field of extrusion and molding, their outstanding use is in the man-made fibre industry. Nylons dominate all other synthetic fibres in quantity and quality and find successful use in fabrics, tire cords, and many industrial fibres.

Polyurethanes. Another important family of synthetic organic polymers is obtained by the reaction of bifunctional isocyanates (containing two $-NCO$ groups) with bifunctional alcohols (two $-OH$ groups). Again, one obtains linear macromolecules that, in this case, are characterized by the regular repetition of the urethane groups ($-O-CO-NH-$). The presence of the urethane groups permits the buildup of a three-dimensional network because of the reaction of the hydrogen of the $-NH-$ group with the terminal isocyanate group of a chain. These networks are either elastomeric or rigid; they are used with great success in the field of elastomeric fibres (spandex) and in making both soft and rigid foams.

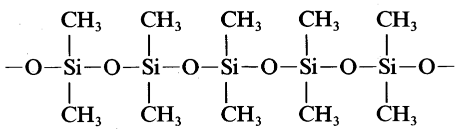
Synthetic polysilicates

Inorganic. Although the synthetic organic polymers are a very large family of useful materials, are the product of a very large industry, and although annual world production is measured in millions of tons, synthetic inorganic polymers have not found any large-scale industrial applications. Synthetic polysilicates (in the form of colloidal particles) are used as fillers and reinforcing agents for plastics used as dentures as well as in synthetic rubbers. Certain polyphosphates with medium molecular weight find some use as flameproofing agents.

The only linear inorganic polymer of known structure and molecular weight is made by the interaction of phosphorus pentachloride (PCl_5) and ammonia (NH_3), with the elimination of hydrogen chloride (HCl). The product is known as polyphosphonitrile chloride; the molecular chains, which have a molecular weight of about 100,000, are composed of repeating $PNCl_2$ units. The substance is a yellow, amorphous, soft, rubbery material that dissolves in many organic solvents and, if stretched to a few times its original length, shows a distinct fibre X-ray diagram, an indication of a regular pattern in the structure of the fibres. Because the material is easily hydrolyzed (broken apart with water), it has found little use as a structural elastomer, but it does have potential use as a polymeric flameproofing agent.

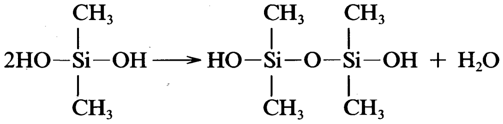
Silicones

Mixed organic–inorganic. Considerable progress was made during the 1950s and 1960s in the synthesis, characterization, and use of linear polymers containing substantial amounts of inorganic elements in the backbone of the chain, together with organic groups attached to the backbone. The most important representatives of this relatively young family of polymers are the polysiloxanes—or silicones, as they are generally called—which have a backbone chain of alternating silicon and oxygen atoms. Also attached to each of the silicon atoms are two monovalent organic groups, such as the methyl ($-CH_3$), or phenyl ($-C_6H_5$) groups. The structure of polydimethyl siloxane is as follows:



$$-\mathrm{O}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{O}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{O}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{O}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{O}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{O}-$$

Polymers of this type are prepared from monomeric dialkylsilanediols by removal of water between molecules. Preparation of the dimer of dimethylsilanediol demonstrates how the process occurs.



$$2\mathrm{HO}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{OH} \longrightarrow \mathrm{HO}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{O}-\underset{\mathrm{CH_3}}{\overset{\mathrm{CH_3}}{\mathrm{Si}}}-\mathrm{OH} + \mathrm{H_2O}$$

Polysiloxanes of low molecular weight, which sometimes have a closed-ring instead of an open-chain structure, are oils and greases. Higher-molecular-weight types are versatile elastomeric materials; they remain soft and rubbery even at temperatures as low as −80° C (−112° F) and, at the same time, are more stable at high temperatures (300° C [570° F] and above) than most conventional plastics and rubbers. Similar polymer chains, in which oxygen atoms alternate with boron, phosphorus, or aluminum atoms, also have been studied, but generally their molecular weights are too low (not higher than 2,000) to provide useful mechanical strength.

STRUCTURE OF POLYMERS

The fundamental consideration in the chemical study of polymers, as of any molecular substance, is the composition of the material in terms of its content of the various kinds of atoms and its structure or arrangement of the atoms in the molecules. Determining the composition is substantially a matter of elemental analysis and does not differ from the procedures used in the study of ordinary inorganic and organic compounds. Furthermore, the composition of the polymer follows simply from the compositions of the monomer, or monomers, used to build it up—with due consideration for such additives as catalysts, activators, modifiers, and stabilizers, some of which may become incorporated in the polymer molecules.

Molecular structure

It is much more complicated, however, to define and, consequently, to establish what is generally called the structure of a polymeric or macromolecular system. The term structure refers, first, to all of the information that can be collected concerning a representative, isolated macromolecule of the material, and second, to the manner in which the individual polymer molecules are geometrically arranged and physically bonded to one another in a macroscopic sample of the material. The former aspect of the structure may be called the molecular, or intramolecular, structure; whereas the latter is the supermolecular, or intermolecular, structure—also referred to as the texture, or morphology.

Molecular weights. In considering the molecular structure, it must always be borne in mind that, with the exception of proteins and certain nucleic acids, all natural and synthetic polymers are heteromolecular, or polymolecular, in character—*i.e.,* they do not consist of a single species of large molecules, but rather of a mixture of many species, which have different molecular weights but substantially the same chemical composition. The best characterization of a given polymer is, therefore, its molecular-weight distribution, which is a representation of the percentages of each molecular weight species in the sample. Determining such a distribution requires a delicate and, sometimes, difficult study of the system; this can always be achieved in a satisfactory manner, however, as long as it is possible to prepare a dilute solution (0.1–2.0 percent) of the polymer in an appropriate solvent and to reprecipitate the polymer from the solution without chemical changes. The procedures for establishing molecular weight distributions can be divided into analytical and preparative methods, depending upon whether one obtains information only about the distribution or whether one actually obtains the individual fractions with different molecular weights.

Methods of analysis

The oldest of the analytical methods is ultracentrifugation, a high-speed centrifugation process that separates molecules in solution according to their molecular weights. Analysis of the distribution of the polymer molecules after centrifugation makes it possible to arrive at a complete description of the spread of molecular weights in the original polymer material. Refinements of this method have increased its applicability to complicated polymeric systems and may make it possible to accomplish a physical separation of the different molecular species in this way.

Gel-permeation chromatography—*i.e.,* separation of molecular mixtures by filtration of a solution through a jellylike material—is a more recently discovered practical way to arrive at the molecular-weight distribution of soluble polymers and may, with further improvements, make it possible to obtain the individual fractions.

Extraction and precipitation are the standard preparation methods, which actually furnish more or less uniform fractions of a polymer, each of which is composed of a narrower molecular-weight distribution than is the original material. With the aid of repeated, careful extraction or precipitation steps or both, it is possible to obtain relatively large amounts of materials that are substantially more uniform than the raw polymer.

The systematic application of these methods has shown that all natural and synthetic polymers, such as cellulose, rubber, polyethylene, nylon, and polyvinyl chloride, have rather broad distribution functions; *i.e.,* they represent mixtures of molecules that differ considerably in their molecular weight. In nylon, for example, the 10 percent of the molecules with largest molecular weight have on the average about 10 times the molecular weight of the 10 percent of lowest molecular weight. For many purposes, it is satisfactory to approximate the molecular-weight distribution by a numerical average: doing this means that a mathematical function is replaced by a single number.

Number and weight averages

In statistical approaches to any subject it is customary to use different averages, which are characterized by the type of quantity averaged. If a polymer is divided into a number of fractions on the basis of molecular weight, two kinds of average molecular weight can be obtained: number-average and weight-average molecular weights. The number-average molecular weight is obtained by adding together the molecular weights of each fraction (*i.e.,* the number of molecules in each fraction times the molecular weight of that fraction) and dividing the total by the number of all molecules in the system. The number-average molecular weight can be measured experimentally by end-group analysis (detection of the ends of the polymer chains) or by measurements of the vapour pressure or osmotic pressure of solutions of the polymer, both effects depending on the numbers of molecules in the solution.

On the other hand, the weight-average molecular weight is obtained by adding together the actual weights of all molecular weight fractions (*i.e.,* the weight of each fraction times its molecular weight) and dividing this sum by the total weight of all molecules in the system. The weight-average molecular weight can be experimentally determined by light scattering, ultracentrifugation, or viscosity measurements of very dilute solutions, all of which are influenced by the weight of material present in the sample.

Separate determinations of the number- and weight-average molecular weights give an idea of the relative uniformity, or nonuniformity, of a given polymer sample. In an ideal case, in which all molecules have a single molecular weight, the number and weight averages are the same. When the polymer sample contains many molecular classes, however, each with a different molecular weight, the two averages are likely to be quite different. When experimental studies show that the number- and weight-average molecular weights are very similar, the polymer itself is relatively uniform—that is, its molecules all have about the same molecular weights. On the other hand, when the two averages are found to be quite different, then the polymer sample has a broad distribution function and the material is polymolecular and nonuniform.

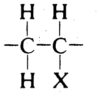
Arrangement of monomer units. The second step in the description of the molecular structure of a polymeric material is the determination of the overall shape of the polymer molecules, or the arrangement of the individual units in the polymer. Clearly, the polymers with the simplest structures are the linear polymers. In such macromolecules the individual structural units, or monomers, are joined to one another in linear sequence. Each terminal unit is connected to only one other unit; each chain is connected to two others. At no point in the chain is any unit connected with three or more other units. Linear macromolecules invariably have only two end groups. Systems composed solely of polymers of this kind show reversible melting and solubility behaviour.

Linear and nonlinear systems

Nonlinear polymeric systems, such as branched macromolecules in which certain of the monomer units are connected with three or four neighbours, are also known. If the number of monomers joined to multiple neighbours is small, the shape of the molecule is essentially that of a linear backbone carrying a number of relatively short branches. In a system of this kind, the individual branched molecules can still be separated from one another without breaking chemical bonds and, as a consequence, materials consisting of branched macromolecules in most cases melt and dissolve reversibly. If, however, the number of multifunctional units in the individual chains becomes relatively large, each develops into a branch and the branches of different backbone chains join together to create a three-dimensional network in which backbone chains are no longer distinguishable from branches and in which all existing units are joined to a number of others by chemical bonds. Systems of this type are called polymeric networks, or cross-linked polymers. Materials in which this type of molecular bonding is preponderant are nonmelting and insoluble, although they may soften and swell to a certain extent when heated or exposed to certain solvents.

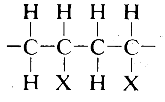
The third step in the description of the molecular structure of a polymeric material is the elucidation of the chain structure or the method of joining of the individual units, or monomers, within the chain.

Depending on the chemical character of the repeating unit, many different types of chain formation are possible even in the case of a linear chain.



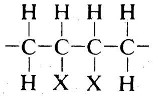
$-\overset{\text{H}}{\underset{\text{H}}{\text{C}}}-\overset{\text{H}}{\underset{\text{X}}{\text{C}}}-$

(1)



$-\overset{\text{H}}{\underset{\text{H}}{\text{C}}}-\overset{\text{H}}{\underset{\text{X}}{\text{C}}}-\overset{\text{H}}{\underset{\text{H}}{\text{C}}}-\overset{\text{H}}{\underset{\text{X}}{\text{C}}}-$

(2)



$-\overset{\text{H}}{\underset{\text{H}}{\text{C}}}-\overset{\text{H}}{\underset{\text{X}}{\text{C}}}-\overset{\text{H}}{\underset{\text{X}}{\text{C}}}-\overset{\text{H}}{\underset{\text{H}}{\text{C}}}-$

(3)

A repeating unit of structure (1) (in which X is any constituent) may, for instance, be joined in two ways to give the structures (2) or (3). Structure (2) represents a "head-to-tail" juncture of the two units, whereas (3) is a "head-to-head" union. Most of the better known vinyltype polymers have molecular chains composed substantially of units joined in head-to-tail fashion. It also is possible, however, for polymer chains to be composed exclusively or in part of units joined head-to-head.

Spatial arrangement in single carbon chains

In a single chain of carbon atoms joined by single bonds, each carbon atom that carries a substituent X is asymmetric—that is, it can exist in either of two forms, or configurations, termed *d* and *l,* which are mirror images of one another (like right and left hand gloves). The various forms of polymer chains that exist depend on the relative configurations of the asymmetric carbons in the chains. If all carbons in a head-to-tail chain, as in structure (2), have the same configuration, the polymer is called isotactic; if the configuration alternates regularly along the chain from *d* to *l,* the molecule is called syndiotactic; and finally, if there is random alternation between the two possible configurations, it is called atactic. (For a discussion of tacticity see below *Classification and nomenclature.*)

Although the structural unit (1) is derived from a particularly simple type of monomer, $H_2C{=}CHX$, it leads to the various possible structures just described. Even more possibilities arise from such monomers as $H_2C{=}CX-CH{=}CH_2$ or $XHC{=}CHY$. Polymer stereochemistry, the discipline that deals with these subjects, is of considerable practical importance, because polymers of different chain structures have distinctly different properties, even when they are derived from the same monomer.

Polymer morphology. Experience also has shown that the ultimate, practical properties of all polymeric materials depend not only on their molecular design—size, shape, and chain structure—but also to a considerable extent on the supermolecular character of the system. This quality, also known as texture, or morphology, of the polymer, is influenced chiefly by the fact that all macromolecules can assume a large number of conformations or spatial shapes, which result from free rotation about single bonds in the chains. These conformations include various helical (spiral) forms, as well as random tangles or coils.

To analyze the supermolecular structure of polymeric systems, it is necessary to focus attention not on any individual macromolecule but on a unit of volume within the investigated object—fibre, film, or molded structure—which contains many macromolecules. In general, a cube with 200-angstrom edges contains many hundreds of polymer molecules and, as a consequence, many thousands of molecular segments. Experience shows that these segments can have several characteristic mutual relationships inside a volume element: (1) they can be randomly distributed in regard to their spatial orientation regardless of whether they belong to the same molecular chain or to different chains; (2) they can possess a degree of lateral (side-by-side) order through the folding of the individual chains; and (3) they can possess lateral order through the parallel arrangement of extended chains.

If all volume elements in the material belong to group (1), the material is an amorphous solid, or glass, which is homogeneous and isotropic (structurally uniform in all directions). It also is noncrystalline and unoriented. Systems of this type usually are obtained when a polymer melt is solidified rapidly by quenching (cooling by sudden immersion).

Orientation and crystallinity

If all volume elements in a sample belong to group (2), the polymer is composed of single crystals, which can be either randomly distributed in space or in a preferred arrangement with respect to a fixed axis or plane. In the first case, the system is unoriented; in the second case, it has a definite orientation. Systems of this type are obtained when a polymer is precipitated slowly from a dilute solution or solidified slowly from a melt.

If all volume elements belong to group (3), the sample consists of extended chain crystallites—that is, crystalline areas or domains—which, again, may be unoriented or may possess particular spatial orientations. Systems of this type are obtained when concentrated solutions or melts are solidified while under the influence of mechanical forces, such as shear or stress.

In practice, any actual polymeric material—in the form of a fibre, film, sheet, or tube—represents a mixture of the three types of volume elements, and its macroscopic character is defined empirically by such expressions as "degree of orientation," "degree of crystallinity," or "later order." It is often difficult to decide whether any given sample has its segments laterally ordered in an intramolecular fashion by means of chain folding, or intermolecularly by way of extended chain crystallization. Extensive experimentation with different methods often is required to discriminate between the two. The degree of orientation is determined quantitatively by optical birefringence (splitting of light waves) or by the measurement of Young's modulus (a measure of elasticity), both properties being dependent on the relative degree of orientation of the chain. The degree of crystallinity, or lateral order, is determined qualitatively by such methods as X-ray and electron diffraction (interference effects observed with X-rays and with beams of electrons) and density, swelling volume, and isotope exchange, all properties influenced by the arrangement of the molecular chains.

Determining the degree of crystallinity

Polymeric materials produced by spinning and related processes, with subsequent mechanical drawing or rolling of the product, consist of more or less oriented mixtures (50–90 percent orientation) of both amorphous and extended-chain crystalline domains (25–85 percent crystallinity). Larger pieces, made by extrusion, injection, or compression, have a typical spherulitic (composed of small spheres) texture containing all three possible segment organizations; the overall degree of orientation is usually low (less than 20 percent), the content of laterally ordered domains is high (up to 95 percent), and chainfolding crystallization is the predominant mode.

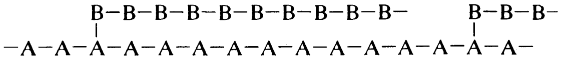
CLASSIFICATION AND NOMENCLATURE

A homopolymer consists of macromolecules formed either by a single type of unit or by two or more chemically different types in regular sequence. Homopolymers may, and frequently do, contain irregularities in minor amounts, especially at the ends of the chains and at junctions leading to branches.

Copolymers are macromolecules containing two or more chemically different monomeric units in a more-or-less irregular sequence. Normally, the relative numbers of the different types of units are not the same in different molecules, nor even in different portions of a single molecule. A mixture of different macromolecules is called a polyblend. A block copolymer is a type of copolymer in which distinct groups of monomeric units of each type are linked together by chemical bonds and then interspersed along the chain. If, for example, the monomer units are designated A and B, a block copolymer of them might have the following structure:

$$-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{B}-\mathrm{B}-\mathrm{B}-\mathrm{B}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-$$

A graft copolymer, on the other hand, contains branches of varying length composed of monomer units of one type joined to a backbone chain made up of monomer units of a second type. Such a copolymer might be composed as follows:



$$-\mathrm{A}-\mathrm{A}-\overset{\overset{\mathrm{B-B-B-B-B-B-B-B-B-B-}}{|}}{\mathrm{A}}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\mathrm{A}-\overset{\overset{\mathrm{B-B-B-}}{|}}{\mathrm{A}}-\mathrm{A}-$$

In their structure and in their properties, graft and block copolymers represent an intermediate form partway between normal copolymers and polyblends.

Tacticity

A tactic polymer consists of macromolecules whose monomeric units follow one another along the chain with their steric configurations (spatial orientations) ordered according to some pattern. The pattern, or tacticity, of the polymer, may consist of one or more simple rules. A eutactic polymer is a tactic polymer whose monomeric units are completely ordered, so that it is devoid of any element of structural disorder. An atactic polymer, on the other hand, is one in whose molecules the spatial configuration of the units is purely random. By way of example, a vinyl polymer prepared by conventional processes is generally atactic; it may possess a regular head-to-tail arrangement, but it is disordered with regard to the steric configurations of the tertiary carbon atoms that follow each other along chains. In isotactic polymers the asymmetric units are ordered in such a way that there is a continual, exact repetition of the unit configuration along the chain. A syndiotactic polymer consists of macromolecules whose monomeric units are asymmetric and ordered in such a way that there is a continual inversion of steric configuration from one monomeric unit to the next down the chain.

Polyelectrolytes are substances that dissociate in water (or other ionizing solvents) to give a polymeric material bearing a great many ionic, or electrically charged, groups, together with an equivalent number of counterions (ions of small charge and opposite sign). Polyelectrolytes can be polyacids (if they release hydrogen ions), polybases (if they release hydroxide ions), polysalts (if they release small cations or anions), or polyampholytes (if they release both hydrogen and hydroxide ions).

Cross-links

Under certain conditions chemical bonds may form between growing polymer chains. Such bonds, called cross-links, form whenever the monomeric units contain functional groups capable of reacting with one another, over and above those consumed in the initial polymerization reaction. The initial result of cross-linking is the formation of small domains, within which the macromolecules consist of random three-dimensional networks of chemically linked chains; such domains are called microgels. If a cross-linking reaction extends over macroscopic dimensions, and if the polymerizing system is taken up by a liquid solvent (or liquid monomer), the result is a gel, or jelly-like material. Drying converts such a gel into an insoluble and nonmelting material, composed entirely of a three-dimensional network at the molecular level. Such thermosetting resins as phenol–formaldehyde and urea–formaldehyde resins and copolymers of styrene and divinylbenzene fall into this category. Similar insoluble and infusible systems are obtained under certain conditions when a vinyl derivative is polymerized in the presence of a small amount of a divinyl compound. The materials thus obtained are called "popcorn" polymers; their exact structure has not been completely elucidated.

POLYMERIZATION PROCESSES

There are many processes by which small molecules can be bonded together to form large ones. Classification of these processes usually is based either on the chemical mechanisms by which the monomeric units are joined together or on the experimental conditions that exist during the reaction. Classifications by reaction mechanism distinguishes between polymerization by addition and polymerization by condensation.

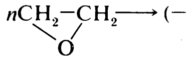
Polymerization mechanisms. *Addition.* Typical cases of addition polymerization are:

Vinyl-type polymerization

1. The addition of small molecules of one type to one another, accomplished by the opening of a multiple bond and without elimination of any part of the molecule, also called vinyl-type polymerization (in the following equations *n* is any number):

$$nH_2C{=}CHX \longrightarrow (-CHX-CH_2-)_n$$

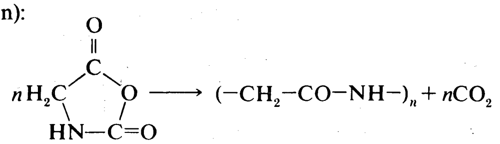
2. The addition of small molecules of one type to one another by the opening of a ring and without elimination of any part of the molecule (epoxide-type polymerization):



3. The addition of small molecules, all of one type, to one another by opening of a multiple bond with elimination of part of the molecule (aliphatic diazo-type polymerization):

$$nRCH = N_2 \longrightarrow (-CHR-)_n + nN_2$$

4. The addition of small molecules of one type to one another by opening of a ring with elimination of part of the molecule (*a*-aminocarboxyanhydride-type polymerization):



5. The addition of small molecules of one type to those of another type by opening of a multiple bond (polyurethane, polyurea, and polyamide formation):

$$nHOCH_2CH_2OH + nOCN(CH_2)_6NCO$$
$$\longrightarrow [-OCH_2CH_2OCONH(CH_2)_6NHCO-]_n$$

6. The addition to one another of biradicals of a single type formed by dehydrogenation (poly-*p*-xylylene polymerization):

$$n(\cdot H_2C-C_6H_4-CH_2\cdot) \longrightarrow (-CH_2-C_6H_4-CH_2-)_n$$

Polyester and polyamide formation

Condensation. A number of possible forms of condensation polymerization are also observed. Typical cases are as follows:

1. The formation of polyesters, polyamides (nylon), and polyethers by elimination of water, or alcohols, from single bifunctional molecules, such as ω-hydroxy- or ω-aminocarboxylic acids or from such combinations as glycols or diamines with diesters or dicarboxylic acids. The reaction of a glycol with a dicarboxylic acid, for example, proceeds as follows:

$$nHO(CH_2)_xOH + nHOOC(CH_2)_yCOOH$$
$$\longrightarrow [-O(CH_2)_xOOC(CH_2)_yCO-]_n + nH_2O$$

2. The formation of polyhydrocarbons by the elimination of halogen or hydrogen halides with the aid of metal and metal halide catalysts:

$$nClH_2C(CH_2)_xCH_2Cl \longrightarrow [-H_2C(CH_2)_xCH_2-]_n + nCl_2$$

3. The formation of polysulfides by the elimination of sodium chloride from bifunctional alkyl or aryl halides with alkali sulfides or by the oxidation of dimercaptans (a type of organic sulfur compound). The former proceeds as follows:

$$nClH_2C(CH_2)_xCH_2Cl + nNa_2S_4$$
$$\longrightarrow [-H_2C(CH_2)_xCH_2S_4-]_n + 2nNaCl$$

Experimental conditions. In actual practice the experimental conditions under which polymerization reactions occur are more commonly used to classify the reactions. Commonly, then, polymerization processes are classified according to whether they occur in the gaseous phase, in the pure liquid phase, or as dispersions, solutions, or emulsions.

Gaseous phase polymerization

1. Polymerization in the gaseous phase may occur at normal, reduced, or elevated pressures. The most important process of this type is the polymerization of ethylene, which is usually carried out at pressures of about 1,500 atmospheres in the temperature range between 190° and 210° C (374° and 410° F). Poly-*p*-xylylene is prepared at reduced pressure by cooling the pyrolyzed monomer from 900° C (1,652° F) to room temperature.

2. Polymerization of single or of several monomer types frequently takes place in the pure liquid phase (bulk polymerization). Many vinyl-type polymerizations are carried out in this manner, particularly when large, transparent pieces (such as plates, rods, lenses) of the final product are desired. Polyester and polyamide formation are carried out almost exclusively in the bulk phase, as is the polycondensation of alkylsilanediols and the formation of phenol–, urea–, and melamine–formaldehyde resins.

3. Polymerization of one or more monomers frequently is carried out by dispersion in the form of small droplets in a liquid that does not dissolve the monomer. (This process also is referred to as suspension, bead, or pearl polymerization.) Styrene, methyl methacrylate, vinyl chloride, and other monomers are polymerized in this manner to give beads of very uniform size and quality for use in certain molding techniques.

4. Polymerization of one or more monomers may take place in solution. If the polymer is soluble in its own

monomer and in the solvent selected for the reaction, the system thickens increasingly as the reaction proceeds. If the polymer is insoluble in the liquid constituents, a slurry forms as the reaction proceeds, and the polymer settles as a more or less swollen powdery mass. Vinyl chloride, acrylonitrile, vinyl acetate, and other monomers usually are polymerized and copolymerized by this technique.

Emulsion polymerization

5. Polymerization of one or more monomers often is conducted in emulsion. It has been found that polymerization frequently takes place in a particularly favourable manner if the monomers are brought together in the form of an aqueous emulsion with the aid of emulsifying agents such as soaps. The polymer or copolymer is then obtained in the form of a latex, or milky liquid, which either can be used as such or can be coagulated by the addition of acids to form agglomerates of the product polymer. Studies of the mechanism of emulsion polymerization indicate that several important steps in the process occur in the micelles (little particles) of the emulsifying agent. The final growth of the macromolecules takes place in polymer particles swollen by monomers transferred by diffusion from the monomer droplets. Most rubbery polymers and copolymers, and some plastics, are produced on a very large scale by emulsion polymerization.

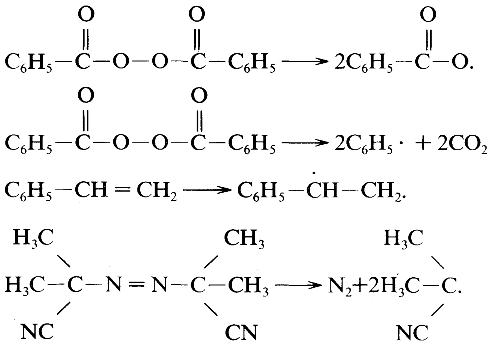
Several other polymerization methods, such as polymerization in the solid state, in a fine powder, or in a foam of small bubble size, have been studied in the laboratory but have not been employed on a large scale for commercial use.

Most polymerization techniques can be carried out in separate batches or as continuous processes. Batch processing still prevails in most large-scale operations, but there is a tendency toward the development of continuous methods for all polymerization processes. Continuous processes have been developed for production of acrylonitrile, nylon, polyester, and certain synthetic rubbers.

Catalysts basic to polymerization reactions

Catalysis. Catalysts, substances that promote chemical reactions without being consumed in the reactions, play an important role in most polymerization reactions. Addition polymerization can be accelerated by three different types of catalysts: free-radical, cationic, and anionic.

Free radical catalysis. Initiation of a polymerization reaction may take place through the initial formation of a free radical, which either is produced by decomposition of a labile (*i.e.,* easily reactive) molecule or is formed in the course of a chemical reaction. Typical cases of free-radical-forming catalysts are peroxides, hydroperoxides, and azo compounds. Various ways of forming free radicals are indicated by equations 1–5:

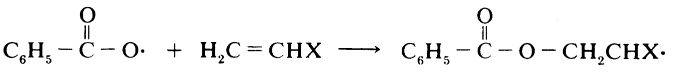


$$C_6H_5-\overset{O}{\overset{\|}{C}}-O-O-\overset{O}{\overset{\|}{C}}-C_6H_5 \longrightarrow 2C_6H_5-\overset{O}{\overset{\|}{C}}-O.$$

$$C_6H_5-\overset{O}{\overset{\|}{C}}-O-O-\overset{O}{\overset{\|}{C}}-C_6H_5 \longrightarrow 2C_6H_5\cdot + 2CO_2$$

$$C_6H_5-CH=CH_2 \longrightarrow C_6H_5-\dot{C}H-CH_2.$$

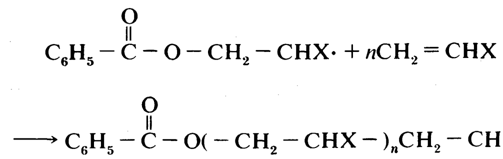
The initiator decomposes either by thermal action at a sufficiently high temperature or at low temperatures under the influence of a photon (packet of light energy) or an electron. The activation energy required to break the critical bond in the molecule of the initiator, usually a O—O, C—N, or C—O bond, amounts to only about 15–30 kilocalories. Some initiators, such as certain hydroperoxides and aliphatic azodinitriles, act at relatively low temperatures (40°–50° C [104°–122° F]); whereas others, such as peroxides or aliphatic azodiesters, require temperatures as high as 150°–175° C (302°–347° F). The free-radical initiator (produced in reactions such as those shown in equations 1–5) attacks the double bond of the monomer, attaches itself to it, and reproduces the free radical at the other end of the adduct:



$$C_6H_5-\overset{O}{\overset{\|}{C}}-O\cdot + H_2C=CHX \longrightarrow C_6H_5-\overset{O}{\overset{\|}{C}}-O-CH_2CHX\cdot$$

The propagation reaction

This process is called the propagation reaction; as it is repeated again and again, it consumes the molecules of the monomer to form macromolecules:



$$C_6H_5-\overset{O}{\overset{\|}{C}}-O-CH_2-CHX\cdot + nCH_2=CHX$$

$$\longrightarrow C_6H_5-\overset{O}{\overset{\|}{C}}-O(-CH_2-CHX-)_nCH_2-CHX\cdot$$

The activation energies of typical propagation reactions have been found to be 5–10 kilocalories.

A free radical at the end of a growing chain can also react with a monomer molecule in such a way that it abstracts a hydrogen atom from the latter and transfers the free radical character from the growing chain to the monomer:

$$R-CH_2-CHX\cdot + CH_2=CHX$$

$$\longrightarrow R-CH_2-CH_2X + CH_2=CX\cdot$$

Such processes are called chain-transfer reactions; they occur between the growing chains and molecules of the monomer or of the already formed polymer. If two free-radical chain ends react with each other, there results a pairing of the lone electrons either by recombination or by disproportionation. Either process leads to the termination of the propagation step and finishes the polymerization reaction by the formation of stable macromolecules. Examples of both are:

$$\ldots CH_2-CHX\cdot + \cdot XCH-CH_2\ldots$$

$$\xrightarrow{\text{recombination}} \ldots CH_2-CHX-CHX-CH_2\ldots$$

$$\xrightarrow{\text{disproportionation}} \ldots CH_2CH_2X + XCH=CH\ldots$$

The interplay of initiation, propagation, and termination reactions leads to average lifetimes for the free radical-type chain ends of 0.1 to 10 seconds and to steady-state free-radical concentrations of 10^{-6}–10^{-8} mole per litre.

To produce free radicals at lower temperatures, such as room temperature or below, the interaction of electron donors, such as iron(II) ions, Fe^{2+}, with electron acceptors, such as hydrogen peroxide, H_2O_2, is utilized. In this way, an oxidation–reduction system is set up, which acts as a permanent source of free radicals:

$$H_2O_2 + Fe^{2+} \longrightarrow Fe^{3+} + HO^- + HO\cdot$$

Such systems are capable of initiating polymerization reactions at temperatures as low as −20° C (−4° F). This reaction activation is of great practical importance in emulsion polymerization, where many undesirable complications can be avoided, or at least suppressed, by working at low temperatures.

Free-radical initiation is employed in the bulk polymerization of styrene and methyl methacrylate, in the solution copolymerization of vinyl chloride–vinyl acetate, in the suspension polymerization of acrylonitrile and vinyl chloride, in the emulsion copolymerization of butadiene–styrene and butadiene–acrylonitrile, and in many other vinyl-type addition polymerizations.

Cationic catalysis. Certain vinyl-type monomers, such as propylene, isobutylene, and most vinyl ethers, do not respond to free-radical initiation but can be initiated by catalysts that produce protons by interaction with a cocatalyst. The protons attack the monomer to form a carbonium ion (a molecular fragment with a positive charge on a carbon atom), which then acts as the carrier for the propagation. A typical example of cationic polymerization

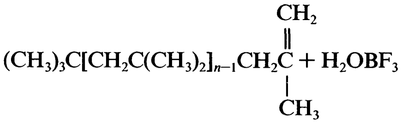
Boron trifluoride as catalyst

is the formation of polyisobutylene from isobutylene under the influence of boron trifluoride as catalyst and water as cocatalyst. The process can be represented by a series of equations:

$$BF_3 + H_2O \rightleftharpoons H_2OBF_3 \rightleftharpoons H^+ + HOBF_3^- \quad (1)$$

$$H^+ + CH_2 = C(CH_3)_2 \longrightarrow (CH_3)_3C^+ \quad (2)$$

$$(CH_3)_3\overset{+}{C} + nCH_2 = C(CH_3)_2 \longrightarrow (CH_3)_3C[CH_2C(CH_3)_2]_{n-1}CH_2\overset{+}{C}(CH_3)_2 \quad (3)$$



$$(CH_3)_3C[CH_2C(CH_3)_2]_{n-1}CH_2\overset{+}{C}(CH_3)_2 + HOBF_3^- \longrightarrow (CH_3)_3C[CH_2C(CH_3)_2]_{n-1}CH_2\underset{\displaystyle CH_3}{\overset{\displaystyle CH_2}{\overset{\|}{\underset{|}{C}}}} + H_2OBF_3 \quad (4)$$

The interaction of catalyst and cocatalyst produces protons (1), the concentration of which is controlled by an equilibrium that depends upon the concentration of the reactants and on the temperature. Other catalysts of this type are aluminum chloride and stannic chloride; other cocatalysts are ethers and amines. The proton adds to the monomer to form a carbonium ion (2), which propagates the chain by subsequent addition to other monomers (3). No formation of free radicals by splitting of an electron pair is necessary; instead, the displacement of an electron pair within the interacting particles suffices to bring about the chemical changes required to combine the monomeric units into macromolecules. As a consequence, the activation energies for initiation and propagation by cationic initiation are much lower than by the free-radical mechanism, and cationic polymerization leads to rapid formation of large molecules even at temperatures as low as −80° C (−112° F).

The termination of cationic polymerization takes place by the reaction of the negatively charged complex ion with the carbonium-ion chain end, a reaction that results in the formation of a terminal hydroxyl group or a terminal double bond (4). Essentially this reaction restores the molecule of the catalyst to its original state so that it can attack another monomer and begin another reaction cycle.

Anionic catalysis. Anions, or negative ions, such as hydroxide (OH^-), amide ion ($NH^-{}_2$, and various carbanions designated R^-, can also initiate vinyl-type addition polymerization under certain conditions. The reaction proceeds by conversion of the monomer to a carbanion, which then undergoes propagation reactions with further monomers. This mechanism is of great importance, and it has been established that most polymerizations initiated by alkali metals or alkali-metal alkyls are of this type. In certain cases the lifetime of the carbanions is very long, so that there is essentially no termination reaction. Polymers formed under such conditions are called living polymers.

Living polymers

Most addition polymerizations involving ring opening are subject to catalytic influence by positively or negatively charged ions. Thus the polymerization of ethylene oxide and other epoxides is strongly accelerated by bases, and the polymerization of α-aminocarboxyanhydrides by acids. Both types are initiated by water, alcohols, or amines.

Coordination-complex catalysis. A very important group of ionic initiators is formed when compounds of transition metals—such as titanium, zirconium, and vanadium—are mixed with metal alkyls—butyllithium, $LiCH_2CH_2CH_2CH_3$, diethylzinc, $Zn(CH_3CH_2)_2$, and triethylaluminum $Al(CH_3CH_2)_3$, for example. In such cases the result is a solid precipitate, that is able to polymerize ethylenic compounds by a suspension type process. Polymerizations of this type take place very rapidly under mild conditions and lead to many products of great importance; *e.g.,* linear polyethylene, isotactic polypropylene, *cis*-1,4-polybutadiene, and ethylene–propylene copolymers.

Condensation catalysis. Most polycondensation reactions also are influenced by acids and bases. In these cases it is important to distinguish between equilibrium and rate effects. Certain materials, such as catalysts, influence the rate, or speed, of chemical reactions without affecting the relative amount of product formed; other substances affect primarily the amount of product present at equilibrium. With many condensation reactions, water-binding agents, such as phosphoric anhydride, sulfuric acid, or trifluoroacetic anhydride, very markedly increase the overall formation of polymer, but they do so by shifting the equilibrium in the direction of polycondensation rather than by a catalytic influence on the elementary process of chain propagation. As most polycondensation reactions are carried out at rather elevated temperatures (200° C [392° F] or higher), it is necessary to employ acids and bases of very low volatility, such as antimony(III) oxide, Sb_2O_3, lead dioxide, PbO_2, calcium hydride, CaH_2, or potassium hydroxide, KOH.

PROPERTIES AND USES OF POLYMERS

Mechanical strength of polymers. The most important property of organic polymers, as compared with ordinary organic substances, is that they exhibit considerable mechanical strength in the solid state. Hence polymers are used to make fibres, films, plastics, rubbers, coatings, and adhesives of commercially valuable properties. The degree to which favourable mechanical properties are exhibited depends upon several factors, the interaction of which determines the most appropriate use of a given polymer and characterizes it as a typical material to form a fibre, plastic, or coating.

Degree of polymerization. One decisive factor for mechanical resistance is the degree of polymerization, or the molecular weight, of the material under consideration. Extensive experiments with many polymers have shown that for each polymer there exists a certain critical molecular weight below which the material shows only poor mechanical strength. Above this critical value, mechanical strength is rapidly developed with increasing molecular weight up to a second level, above which the increase of mechanical resistance with increasing molecular weight becomes less pronounced. An industrially useful material must have, therefore, a certain molecular-weight value above which only a diminishing return of strength is gained. As a matter of safety, it is advisable to choose a molecular weight well above this point, but it has been found that materials with too high a degree of polymerization exhibit exorbitant viscosities in solution or in the molten state, which render such important processes as filtration, spinning, casting, and molding unduly difficult. A proper compromise between easy processing ability and desirable ultimate properties must therefore be established by choosing an optimum molecular-weight range in each individual case.

Optimum molecular weight

Intermolecular forces. Another important factor for the development of characteristic mechanical properties is the magnitude of the intermolecular forces between individual macromolecules in a given material, as well as the structural regularity of the macromolecules themselves. If the intermolecular attraction is small and the chains are difficult to fit laterally into a lattice-like arrangement, then at a particular temperature and stress material that has been stretched shows a distinct tendency to return to a relaxed state in which the macromolecules are again curled rather than laterally ordered. Such behaviour is typical of rubbers (elastomers). On the other hand, if the molecular cohesion is strong and the fine structure of the chains provides for easy lateral fitting into lattice-like aggregates, a state of substantial crystallinity is favoured and the material is a typical fibre. In intermediate cases, in which the forces are moderate and the geometry of the chains is not very favourable for crystallization, the behaviour of the material will depend greatly upon external conditions such as temperature and pressure. Such systems are typical plastics.

These qualitative considerations show that rubbers, plastics, and fibres are not intrinsically different materials. Their differences are rather a matter of degree and result from the combined effect of the intermolecular forces between the long chains and the general tendency of the chains themselves to curl and fold, which leads to a mixture of crystallized and disordered portions in a given sample of the material. In general, it can be said that under given conditions, such as temperature and external forces applied to the sample, the tendency to crystallize is brought about by two different factors, which to a certain extent conflict with each other:

Intermolecular forces versus molecular geometry

1. Forces between the individual macromolecules: if these forces are strong, then at ordinary temperatures they chiefly determine the behaviour of the material, and it can be expected to behave like a hard plastic or like a typical fibre. This condition appears to be true for cellulose, certain proteins, nylon, polyesters, and polyacrylics.
2. The microgeometry of the molecular chains themselves: If the chains fit easily into a crystal-like lattice, as polyethylene, nylon, and certain other substances do, the material crystallizes under the influence of comparatively weak forces and hence, in general, has a tendency to appear more fibre-like than would be expected if the intermolecular forces alone were considered. On the other hand, materials with bulky and irregular macromolecular chains do not crystallize even if the intermolecular forces are quite strong. These are soft plastics and elastomers, and they behave more like rubber than would be expected on the basis of the intermolecular forces alone.

Commercially important polymers. From the practical point of view, classification and application of organic polymers is dominated by the raw materials (monomers) from which they are derived, because the monomeric units are responsible not only for the intermolecular forces and the structure of the individual polymer molecules but, in most cases, also for the cost of the final product.

Polyhydrocarbons. Chemically the simplest class of commercial polymers is the polyhydrocarbons. These are very widely used and cover the entire range from very soft and resilient rubbers, through plastics of all kinds, to strong, tough, and high-melting fibres.

Uses of polyethylene, polypropylene, and polystyrene

The most important commercial hydrocarbon polymer is polyethylene, which exists in two types: a lower density material, which is made by high-pressure polymerization, and a high-density type, which is made with the aid of special catalysts at normal pressure. Both are processed by various molding, extrusion, and casting techniques and are widely used to produce such products as trash cans, pails, containers for milk and most household liquids, shopping bags, and wrapping films for many dry goods. Polypropylene is somewhat stiffer and has a higher melting point and a more brilliant and transparent appearance. It is used for the same purposes as polyethylene but also, because of its higher melting point, as a fibre for carpets, upholstery, drapings, and, mixed with wool or acrylic fibres, for blankets. Polystyrene is brilliantly transparent and excellently moldable, but it has a somewhat lower softening point and it is more likely to swell in organic solvents than polyethylene or polypropylene. It is widely used for many household goods, as well as toys, pens, pencils, and decorative molded items. These three polyhydrocarbons all have relatively regular, easily orientable molecular chains, which can be readily brought into laterally ordered bundles resembling a crystal-like structure. They are used, therefore, as plastics and fibres, in spite of the fact that the intermolecular forces between the hydrocarbon chains are low. Other important polyhydrocarbons, however, with more bulky chains, which show more resistance to lateral order, have the properties of a typical rubber. Natural rubber—*cis*-polyisoprene—is the classical representative of these substances. Others are the principle synthetic elastomers, such as polybutadiene and copolymers of butadiene or isoprene with styrene and with each other. It is interesting that random copolymers of ethylene and propylene are also typical elastomers, because the irregular sequence of the two monomers along the chain reduces their capacity for the formation of laterally ordered semicrystalline domains.

Uses of polyvinyl chlorides and polyacrylonitrile

Chlorinated and cyanated polyhydrocarbons. Commercially important classes of polymers are the chlorinated and cyanated polyhydrocarbons. In these, the presence of a chlorine atom or cyano-group increases intermolecular attraction and has the consequence that regularly built chains of these materials behave like fibres or plastics. Outstanding examples are the successful synthetic fibres made of polyvinylidene chloride, polyvinyl chloride, and polyacrylonitrile. If irregularity is introduced into the molecular chains by copolymerization or by the use of bulky monomers, elastomeric materials are obtained. Examples are neoprene—a polychloroprene rubber—and the copolymers of butadiene with acrylonitrile, which form a large and important class of synthetic elastomers.

Polyesters. Intermolecular forces between adjacent macromolecular chains can also be increased by the introduction of an ester bond into the backbone of the chain; this addition leads to the large and important family of polyesters. If the chains are regularly built and easy to align, typical fibre-forming polymers are obtained, such as polyethylene glycol terephthalate, which is today one of the most successful synthetic fibres. Polyesters with less regular chains form a large variety of plastics that are reversibly softening and behave like typical thermoplastics; such polyesters can be cross-linked and thereby made into hard, insoluble, and nonmelting thermoset materials.

Uses of nylons

Polyamides. Still larger intermolecular forces are obtained when an amide bond is introduced into the backbone of linear macromolecules, producing the large and important family of polyamides, or nylons. These have the distinction of being superior fibre-forming materials, and in the late 1930s they initiated the introduction of synthetic fibres into the textile industry. Polyamides with reduced chain regularity produce excellent plastics, which are used in molding and extrusion processes. Polyacrylic compositions excel by their weather resistance, scratch resistance, high gloss, and capacity to accommodate pigments and other fillers. As a consequence, they represent an extremely successful class of materials for synthetic paints and coatings, widely used on products ranging from furniture to automobiles and airplanes.

Other polymers. Other polymers of some practical importance are: the polyurethanes as rubbers, foams, and adhesives; the polyfluorocarbons and the silicones because of their exceptional chemical inertness; the epoxy resins as coatings and adhesives; and several linear macromolecules composed of aromatic rings, because of their useful behaviour at temperatures as high as 600° C (1,112° F).

(H.F.M.)

PRINCIPLES OF CHEMICAL SYNTHESIS

Chemical synthesis is concerned with the construction of complex chemical compounds from simpler ones. Several million compounds have been synthesized, and their number is being added to at the rate of some hundreds of thousands every year.

GENERAL CONSIDERATIONS

A synthesis usually is undertaken for one of three reasons: first, to meet an industrial demand for a product. For example, ammonia is synthesized from nitrogen and hydrogen and used to make, among other things, ammonium sulfate, employed as a fertilizer; vinyl chloride is made from ethylene and used in the production of polyvinyl chloride (PVC) plastic. In general, a vast range of chemical compounds is synthesized for applications as fibres and plastics, pharmaceuticals, dyestuffs, herbicides, insecticides, and other products.

Second, an enormous number of compounds of considerable molecular complexity occur naturally, in both living organisms and their degradation products; examples are proteins (in animals) and alkaloids (alkaline materials found in plants). The syntheses of these natural products have usually been undertaken in the context of the determination of the structures of the compounds; if a material is deduced to have a particular structure on the basis of its chemical reactions and physical properties, then the finding that a compound synthesized by an unambiguous method for this structure is identical to the natural product provides confirmation of the validity of the assigned structure.

Third, a synthesis may be carried out to obtain a compound of specific structure that does not occur naturally and has not previously been made, in order to examine the properties of the compound and thereby test theories of chemical structure and reactivity.

GENERAL APPROACH TO SYNTHESIS

There is, then, an essentially limitless range of compounds that are capable of being synthesized. In practice, the synthesis of a preselected compound is made possible by particular chemical groupings undergoing transformations that, while they are dependent upon the conditions applied to the compound, are largely independent of the structure of the remaining part of the molecule. Thus, the combination of a knowledge of the structure of the compound to be synthesized and the general types of transformation that compounds undergo enables a synthesis to be planned. The general approach, cut to its barest essentials, is to examine the structure of the desired end product—for example, *Z*—and to deduce the structure of some (slightly simpler) compound—for example, *Y*—that should be capable of transformation into *Z* by a reaction of known type. A possible precursor of *Y* is sought in similar manner, and in this way the chain of compounds is extended until a compound, *A*, is reached that is available for the work; the necessary transformations, beginning with *A* and ending with *Z*, are then carried out. Most individual steps in the sequence result in a change in only one bond; some result in changes in two bonds at a time, but it is rare for more extensive changes to occur.

Evaluation of a synthetic method. Three factors must be borne in mind when evaluating a particular synthetic plan. The first is cost, of far greater importance in industrial, large-scale synthesis than in laboratory work in which a particular synthesis may be carried out only once, as in the total synthesis of a naturally occurring compound, and which in any case is likely to be on a relatively small scale.

Importance of yield

Second, the yield in each step must be considered. A step in a synthesis may be theoretically possible but in practice almost impossible to achieve. For example, a proportion of the reactant may be converted into a different product by an alternative process that competes with the desired one, some of the product may undergo a subsequent reaction, or some of the product may be lost in the separation processes required for its isolation in a pure state. The yield is usually defined, on a percentage basis, as the number of molecules of product obtained when 100 could in principle have been formed. A yield of about 80 percent or more is generally considered good, but some transformations can prove so difficult to achieve that even a yield of 10 or 20 percent may have to be accepted.

Naturally, the yield of a process affects the cost of the product, because the shortfall from a 100 percent yield represents wasted material. In addition, yield can be of the utmost importance in determining whether or not a synthesis is a practicable possibility, because the overall yield of a synthesis is the product of the yields of the individual steps. If these intermediate yields are mostly low, the ultimate product may not be obtainable in the necessary amount from the available starting material.

Finally, consideration must be given to the rate at which each step in the planned sequence occurs. In many instances, a desired reaction is possible in principle but in practice takes place so slowly as to be ineffective. It is then necessary to investigate whether the rate can be increased to a practicable level by altering the conditions of the reaction—for example, by raising the temperature or by adding an extra species that increases the rate without altering the nature of the reaction.

The factors that have been briefly mentioned above as important in determining the course of a synthesis will now be considered in detail.

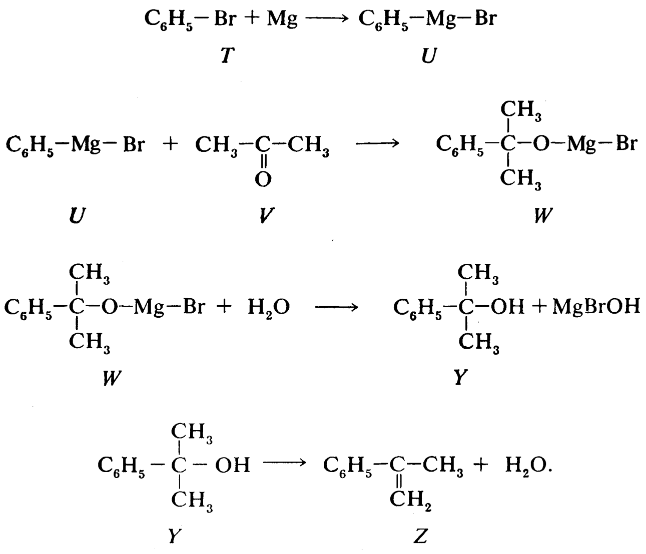
Sources of starting materials

Availability of starting materials. All syntheses must be planned so as to start with available materials. A few elements occur naturally and require only a simple separation process for their isolation; examples are oxygen and nitrogen in the atmosphere and the metals gold and silver. Most metals occur, however, as minerals, in combined form, and an appropriate chemical extraction process must be applied if the element itself is required.

The main source of organic compounds is now petroleum, which has superseded coal. Petroleum consists mainly of straight-chain alkanes (hydrocarbons of general formula C_nH_{2n+2}, such as methane, CH_4, ethane, CH_3CH_3, propane, $CH_3CH_2CH_3$, etc.) with up to about 125 carbon atoms in the chain, together with smaller quantities of cyclic alkanes and aromatic hydrocarbons such as benzene; the proportions vary from one oil field to another. The mixture is separated by fractional distillation, based on the fact that the compounds in the mixture have different boiling points. From the resulting compounds, others are formed by chemical processes; for example, the compounds containing 16–20 carbon atoms are passed through a catalyst, a fine powder made of silica and aluminum oxide, and heated at 450°–500° C (850°–950° F), resulting in the formation of smaller compounds including olefins (compounds containing the unit $>C{=}C<$) such as ethylene, a process called catalytic cracking. From such processes a large number of compounds is available on an industrial scale, at prices that range upward from a few cents per pound.

Planning the synthesis. *Possible synthetic routes.* As has been noted, the usual approach is to work backward in a stepwise fashion until an available starting material is reached.

If, for example, the compound sought, again labelled *Z*, is an olefin, a possible precursor is an alcohol because it is known that olefins can be obtained (among other ways) from alcohols by removal of water (dehydration). One route to an alcohol is by reaction of one of a group of compounds called Grignard reagents with a compound containing the carbonyl group $>C{=}O$; an appropriate combination in this case would be the compounds *U* + *V*, the Grignard reagent *U* being available from the compound *T*. As the materials *T* and *V* are available in most laboratories, one possible scheme for synthesizing *Z* is now complete:



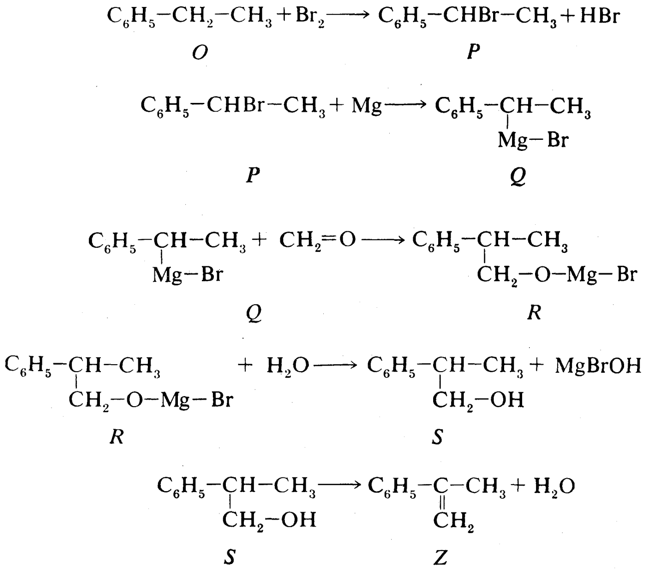
$$\underset{T}{C_6H_5{-}Br} + Mg \longrightarrow \underset{U}{C_6H_5{-}Mg{-}Br}$$

$$\underset{U}{C_6H_5{-}Mg{-}Br} + \underset{V}{CH_3{-}\underset{\underset{O}{\|}}{C}{-}CH_3} \longrightarrow \underset{W}{C_6H_5{-}\overset{\overset{CH_3}{|}}{\underset{\underset{CH_3}{|}}{C}}{-}O{-}Mg{-}Br}$$

$$\underset{W}{C_6H_5{-}\overset{\overset{CH_3}{|}}{\underset{\underset{CH_3}{|}}{C}}{-}O{-}Mg{-}Br} + H_2O \longrightarrow \underset{Y}{C_6H_5{-}\overset{\overset{CH_3}{|}}{\underset{\underset{CH_3}{|}}{C}}{-}OH} + MgBrOH$$

$$\underset{Y}{C_6H_5-\overset{\overset{CH_3}{|}}{\underset{\underset{CH_3}{|}}{C}}-OH} \longrightarrow \underset{Z}{C_6H_5{-}\underset{\underset{CH_2}{\|}}{C}{-}CH_3} + H_2O.$$

Other possibilities would certainly be considered. First, there is an alternative alcohol, *S*, from which the product could be derived; working backward from *S*, by way of the appropriate Grignard reagent, gives a possible route from *P*. *P* is unlikely to be available in the laboratory, however; if it is not, it requires synthesis, for example, from *O*. At this stage it is realized that the planned synthesis is longer than that from *T* and *V*, and, unless some specific advantage can be seen—for example, much better yields, which are unlikely in this case—it is discarded.



$$\underset{O}{C_6H_5{-}CH_2{-}CH_3} + Br_2 \longrightarrow \underset{P}{C_6H_5{-}CHBr{-}CH_3} + HBr$$

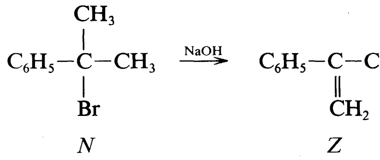
$$\underset{P}{C_6H_5{-}CHBr{-}CH_3} + Mg \longrightarrow \underset{Q}{C_6H_5{-}\underset{\substack{|\\ Mg{-}Br}}{CH}{-}CH_3}$$

$$\underset{Q}{C_6H_5{-}\underset{\substack{|\\ Mg{-}Br}}{CH}{-}CH_3} + CH_2{=}O \longrightarrow \underset{R}{C_6H_5{-}\underset{\substack{|\\ CH_2{-}O{-}Mg{-}Br}}{CH}{-}CH_3}$$

$$\underset{R}{C_6H_5{-}\underset{\substack{|\\ CH_2{-}O{-}Mg{-}Br}}{CH}{-}CH_3} + H_2O \longrightarrow \underset{S}{C_6H_5{-}\underset{\substack{|\\ CH_2{-}OH}}{CH}{-}CH_3} + MgBrOH$$

$$\underset{S}{C_6H_5{-}\underset{\substack{|\\ CH_2{-}OH}}{CH}{-}CH_3} \longrightarrow \underset{Z}{C_6H_5{-}\underset{\substack{\|\\ CH_2}}{C}{-}CH_3} + H_2O$$

Second, a precursor to *Z* other than the alcohols *Y* and *S* would be considered. For example, a possibility is *N*. The most obvious route to *N*, however, is from the alcohol *Y*, so that this method would simply involve an extra, unnecessary step compared with the route involving the conversion of *Y* into *Z* directly.



$$\underset{N}{C_6H_5{-}\overset{\substack{CH_3\\ |}}{\underset{\substack{|\\ Br}}{C}}{-}CH_3} \xrightarrow{NaOH} \underset{Z}{C_6H_5{-}\underset{\substack{\|\\ CH_2}}{C}{-}CH_3}$$

Evaluation of alternatives. Once a scheme with the minimum number of transformations in it has been selected, attention must be given to three other questions for each of the individual steps.

The first of these is whether the step will succeed as planned in this particular case. For example, reaction with the Grignard reagent was chosen because it is known that normally a Grignard reagent reacts with a compound containing the carbonyl group to give an alcohol. Some exceptions to this are also known, however; if, for example, the carbon atom in the carbonyl group is attached to two very bulky groups, the Grignard reagent can be prevented from reaction. Therefore, an examination must be made of any grounds for suspecting that the reaction might fail.

The second question is whether any unwanted transformation will occur at the same time. It frequently happens that there are two or more positions in a compound with which a particular agent can react. In such cases, unless that position at which reaction is desired is much more reactive than the others, it is necessary to modify the other positions to prevent them from reacting. This technique, called protection, is described in more detail later.

The third question is whether the operations required are practicable in the laboratory. For example, some reactions require very high pressures, and the necessary equipment is not widely available; others involve hazards that can be adequately dealt with only in highly specialized laboratories, as in the case of radioactive materials.

The problems in planning a synthesis increase rapidly with the degree of complexity of the target molecule for two reasons: first, the number of possible combinations of steps that must be considered in reaching a decision as to the best route quickly multiplies; for example, if there were two ways in which the final step could be effected, two for the preceding step, and so on, then the number of possibilities to consider is 32 for a five-step synthesis, over 1,000 for a 10-step synthesis, and so on. Second, as the molecule becomes more complex, the possibility of finding a reagent that will effect a desired transformation without bringing about unwanted processes is reduced. Even the most thoughtfully conceived scheme can founder because there was no hint from any previous observation that a particular step would fail as a result of a combination of factors specific to the target molecule in question and never previously encountered. It can be readily appreciated that some of the more complex compounds that have been synthesized—for example, the steroid cholesterol, the alkaloid strychnine, and the plant material chlorophyll—have each taken many man-years of highly skilled effort for their achievement. The failure of a step in a synthesis, perhaps formation of a product different from the one intended, can sometimes be turned to advantage in revealing a new aspect of reaction mechanism that can be applied in later syntheses.

The protection of reactive positions. A plan to effect reaction at a particular position in a compound with a specific reagent can be thwarted if in the molecule there is another position or group that preferentially reacts with that reagent. In such cases it is sometimes possible to block this more reactive position temporarily, while the planned reaction is carried out. The technique is to introduce another reagent that converts the more reactive grouping into a derivative that is inert to the desired reagent; the latter can then be introduced so as to react at the intended position, and finally a third reaction is used to reconvert the blocked position to its original form.

Factors affecting reaction rate. Some reactions take place instantaneously at room temperature simply on mixing the reagents. This is especially true of reactions between inorganic salts; for example, when an aqueous solution of silver nitrate is mixed with one of sodium chloride, silver chloride immediately precipitates from the solution. The underlying reason for this rapidity is that these compounds both contain ions that move freely in solution and collide frequently with each other; collision between a silver ion and a chloride ion can result in the two particles staying together and gradually coagulating with other such pairs as a precipitate, because silver chloride is insoluble in water.

Reactions of covalent compounds

In contrast, the reactions of covalently bonded compounds usually occur slowly at room temperature. The reason is that one or more of these bonds generally has to be broken, and energy is required for that process; the energy is normally supplied in the form of heat but in some circumstances can be supplied as visible or ultraviolet light. The reaction of iodomethane, CH_3I, with sodium hydroxide, NaOH, in solution, to give methanol, CH_3OH, and sodium iodide, NaI, is exemplary. Sodium hydroxide is an ionic compound containing Na^+ and OH^- ions, and its reaction with iodomethane consists of the formation of a covalent bond between the oxygen atom of OH^- and the carbon atom of CH_3I while at the same time the C—I bond breaks, the two electrons in the bond remaining with the iodine atom, which consequently becomes an iodide ion. The first requirement for reaction to occur is a collision between the OH^- ion and CH_3I, but it is also necessary that the molecules possess enough kinetic energy on collision in order to supply the amount necessary for bond breaking. At room temperature only a small proportion possess the required amount of energy, but as the temperature is raised the average amount of kinetic energy of the particles increases and, with it, the proportion of molecules that possess enough energy for reaction. The critical amount of energy—*i.e.*, the activation energy—varies from reaction to reaction, and so does the dependence of the rate on temperature, but, as a rough guide, for many reactions the rate doubles for each 10° C (18° F) rise in temperature.

Reactions in liquids

The physical state of the reactants also has to be considered. Two solids do not react at a significant rate because when they are put in contact the rate at which the molecules diffuse from one into the other and thereby collide with each other is negligible. In both gases and liquids, however, collisions occur millions or billions of times a second; consequently, a solid reactant is usually brought into the liquid state by heating, or, if its melting point is undesirably high, it is dissolved in a solvent. A commonly employed method is to dissolve the reactants in a solvent, such as water or benzene, and then to heat the solution at its boiling point.

Reactions in the gas phase, which are mostly of industrial concern, can be carried out in temperature-controlled furnaces. For example, in the Haber process for the manufacture of ammonia, a mixture of nitrogen and hydrogen is heated at a pressure of 300 atmospheres and a temperature of about 500° C (950° F).

Catalysis. Many reactions occur too slowly even at elevated temperatures to be practicable procedures, but their rates can sometimes be enormously increased by the addition of a catalyst (defined broadly as a material that speeds a reaction but is unchanged at the end of it).

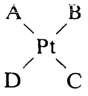
Catalysts can be divided into two categories—heterogeneous and homogeneous; the former group are solids suspended in a solution of the reactants or over which gaseous reactants are passed, and the latter are materials that are soluble in the reaction medium.

A simplified interpretation of the ability of a heterogeneous catalyst, usually a metal, to increase the rate of a reaction is that molecules of each reactant become adsorbed on its surface in close proximity, and their internal bonds become weakened somewhat so that reaction is facilitated. Homogeneous catalysts, on the other hand, almost always take part chemically in the reactions they catalyze but are regenerated at the end of the process. Metallic compounds have proved valuable as catalysts for organic reactions, including those of industrial importance. The key to success with these catalysts lies in the transient formation of an intermediate compound.

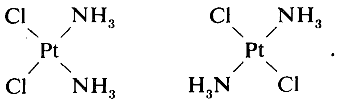
Metallic compounds as catalysts

Activation. Closely akin to catalysis is the activation of a compound by a slight modification of one of its groups.

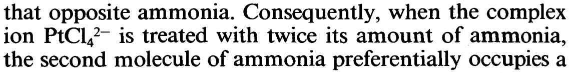
An activating influence of a different type occurs in reactions of inorganic complexes of platinum, and it is useful to consider this behaviour in detail. Platinum exists in salts as the ion Pt^{2+}, usually bonded to four molecules or ions located at the corners of a hypothetical square of which platinum is the centre:



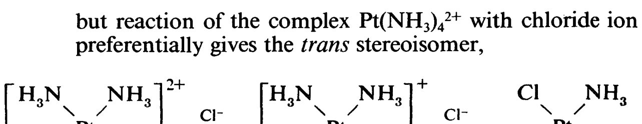
The ease with which a particular group, A, for example, is replaced by another group is much more strongly dependent upon the nature of the group opposite (*trans*) to it, C, than on those adjacent to it, B and D. This *trans* effect can be exploited in synthesis, as, for example, in the preparation of the two stereoisomers of the compound $PtCl_2(NH_3)_2$:



Here, use is made of the fact that a chloride ion, Cl^-, exerts a greater *trans* effect than does ammonia, NH_3; that is, the group *trans* to chloride is more readily replaced than is that opposite ammonia. Consequently, when the complex ion $PtCl_4^{2-}$ is treated with twice its amount of ammonia, the second molecule of ammonia preferentially occupies a position adjacent to the first, giving the *cis* stereoisomer,



but reaction of the complex $Pt(NH_3)_4^{2+}$ with chloride ion preferentially gives the *trans* stereoisomer,



Isolation and purification of products. The product of a synthesis is normally contaminated with reagents used in the synthesis, by-products, and possibly some unchanged starting material; these contaminants must be removed in order for a pure sample to be obtained. In a multistep synthesis, it is normally desirable to purify the product from each step before proceeding to the next. Various techniques for isolation and purification are discussed in the article ANALYSIS AND MEASUREMENT: *Chemical separations and purifications.*

RECENT DEVELOPMENTS IN SYNTHETIC METHODS

The art of chemical synthesis is a constantly evolving one in which new approaches and techniques are developed to meet new challenges. There are two recent developments of a novel kind that are likely to become of increasing importance.

First, it has been shown that synthetic schemes can be efficiently planned with the aid of a computer. The input consists of the various methods for transforming particular types of grouping into others, together with the known information as to the suitability of each process in particular structural situations. For any particular target molecule, the computer can then select various synthetic schemes and can choose between these on the basis of such factors as ease of handling, likely yields, or other considerations.

Automated synthesis

Second, the practicability of the automation of synthesis has been proved for the case of polypeptides. Briefly, polypeptides and proteins are polymers that occur in living systems, and they are best synthesized from amino acids, species that contain both carboxylic acid groups and amino groups. In the laboratory, however, problems arise. First, amino acids do not react together readily; it is necessary to activate the carboxylic acid group of one molecule by converting it into some suitable function that will react readily with the amino group of the second molecule. Second, the activated group is approximately as likely to react with the amino group of another molecule of the same type as with that of a molecule of the second amino acid. To obviate this problem, the amino group of the first molecule is first protected by the application of a suitable reagent. The sequence of operations, then, is to protect one amino acid at its amino end, activate it at its carboxylic acid end, and treat it with the second amino acid; the protecting group is then removed, and a third amino acid, likewise protected at its amino end and activated at its carboxylic end, is introduced; the synthesis continues with repetition of this cycle. The outstanding problem has been that, after each chemical operation, it has been necessary to purify the intermediate product so as to prepare it for the next operation, and this has not only been considerably time-consuming but has also resulted in the loss of material at each step; since dozens or even hundreds of steps are needed, the procedure has proved impracticable for very large molecules. The advance, simple in concept and remarkably effective in practice, has been to attach the first amino acid in the sequence, by means of a reaction generating a covalent bond, to an insoluble solid. The second amino acid, duly protected and activated, is introduced, and, after reaction, soluble reagents and by-products are simply removed by washing the solid, the protecting group is removed by chemical reaction and washing, and the next amino acid is introduced. No purifications other than washing are necessary until the final polypeptide, having been removed from the solid by chemical reaction, is purified. The losses normally sustained during conventional purification methods are thereby avoided, so that yields are high, and each of the individual steps can be carried out in quick succession.

Moreover, apparatus can be designed so that the fairly small number of reagents needed, including each amino acid in its protected and activated form, can be automatically delivered into the reaction vessel at the appropriate time, according to a predetermined program based on the particular polypeptide required. (R.O.C.N./Ed.)

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(W.F.K.)

Inorganic compounds: Comprehensive works: F.A. COTTON and G. WILKSON, *Advanced Inorganic Chemistry: A Comprehensive Text*, 2nd ed. (1966), the standard text; N.V. SIDGWICK, *The Chemical Elements and Their Compounds*, 2 vol. (1950), the classic compilation; C.S.G. PHILLIPS and R.J.P. WILLIAMS, *Inorganic Chemistry*, 2 vol. (1965–66), a sound and extensive survey of the field; P.J. and B. DURRANT, *Introduction to Advanced Inorganic Chemistry*, 2nd ed. (1970), a relatively thorough presentation of the subject; H.J. EMELEUS and A.G. SHARPE (eds.), *Advances in Inorganic Chemistry and Radiochemistry*, vol. 1 (1959–), annual reviews on specific aspects of inorganic chemistry; INTERNATIONAL UNION OF PURE AND APPLIED CHEMISTRY, *Nomenclature of Inorganic Chemistry* (1959), commission report on internationally accepted methods of naming inorganic compounds. See also WILLIAM H. PORTERFIELD, *Inorganic Chemistry: A Unified Approach* (1984); BODIE E. DOUGLAS, DARL H. MCDANIEL, and JOHN J. ALEXANDER, *Concepts and Models of Inorganic Chemistry*, 2nd ed. (1983); DAVID B. RORABACHER and JOHN F ENDICOTT (eds.), *Mechanistic Aspects of Inorganic Reactions* (1982); D.T. BURNS, A. TOWNSEND, and A.G. CATCHPOLE, *Inorganic Reaction Chemistry: Systematic Chemical Separation*, 2 vol. in 3 (1980–81); STEPHEN J. LIPPARD (ed.), *Progress in Inorganic Chemistry* (1983).

(D.H.B.)

Studies on water: Theoretical and experimental investigations of the structure and properties of water and aqueous solutions form the subject of a number of recent books including J.L. KAVANAU, *Water and Solute-Water Interactions* (1964); O.Y. SAMOILOV, *The Structure of Aqueous Electrolyte Solutions and the Hydration of Ions* (1965; orig. pub. in Russian, 1957); and D. EISENBERG and W. KAUZMANN, *The Structure and Properties of Water* (1969). Collections of papers by various authors treating in detail topics related to water include A.K. COVINGTON and P. JONES (eds.), *Hydrogen-Bonded Solvent Systems* (1968); and R.A. HORNE (ed.), *Water and Aqueous Solutions: Structure, Thermodynamics and Transport Processes* (1972). N.E. DORSEY, *Properties of Ordinary Water Substance in All Its Phases* (1940), although old, contains a most complete and detailed compilation of water properties. A number of journals and volume series are devoted in whole or in part to water chemistry, water pollution, water management, and oceanography. These include *Advances in Hydroscience, Water Resources Research, Limnology and Oceanography, Environmental Science and Technology, Geochimica et Cosmochimica Acta, Journal of Geophysical Research, Journal of the Water Pollution Control Federation,* and the *Journal of Solution Chemistry.* In the past much of the fundamental research on water appeared in either the *Journal of Physical Chemistry* or the *Journal of Chemical Physics.* For later research see JON C. VAN LOON, *Chemical Analysis of Inorganic Constituents of Water* (1982); ROLF K. FREIER, *Wasseranalyse: Physiko-chemische Untersuchungsverfahren wichtiger Inhaltsstoffe* (1964); OLOF ALEXANDERSSON, *Living Water* (1982); AMERICAN WATER WORKS ASSOCIATION, *Introduction to Water Quality Analyses* (1982).

(R.A.Ho./Ed.)

Organic compounds: E.H. RODD (ed.), *Chemistry of Carbon Compounds*, 10 vol. (1951–62; 2nd ed. by S. COFFEY, 1964–), is the most extensive general text on organic chemistry in English. *Beilsteins Handbuch der organischen Chemie* is the only source in any language that attempts to cover the chemical and physical properties of all organic compounds with references to the original literature. The fourth edition of 27 volumes published during the years 1918–37 covers all known organic compounds through 1909. Publication of the first supplement of 31 volumes covering the years 1910–19 began in 1928 and was completed in 1938. The second supplement of 31 volumes covers 1920–29. It was published between 1941 and 1957. General indexes cover all of the literature through 1929. The third supplement covering 1930–49, published 1958–73, consists of 49 volumes. Until a general index is published for this supplement, the indexes to each volume must be used. Publication of the fourth supplement, covering 1950–59, began in 1973. *Chemical Abstracts*, published by the AMERICAN CHEMICAL SOCIETY, Washington D.C., is a weekly journal that summarizes world literature in all branches of chemistry. A collective formula index was published covering the years 1920–46. It was followed by supplementary indexes for 1947–56, 1957–61, 1962–66, and 1967–71. For data from 1972 to date, there are semiannual indexes. R.C. WEAST (ed.), *Handbook of Chemistry and Physics*, 63rd ed. (1982–83), lists the common physical properties for more than 15,000 organic compounds. The *Nomenclature of Organic Chemistry*, 2nd ed. (1966), published by the INTERNATIONAL UNION OF PURE AND APPLIED CHEMISTRY, provides the internationally approved rules of nomenclature. The introduction to the index of *Chemical Abstracts*, vol. 56 (1962), discusses the naming of organic compounds followed for indexing purposes. See also R.S. CAHN, *An Introduction to Chemical Nomenclature* (1959); and the UNITED STATES TARIFF COMMISSION, *Synthetic Organic Chemicals* (1970), for production information. For discussion of specific compounds see WILHELM KEIM (ed.), *Catalysis in C_1 Chemistry* (1983); DONALD L. WISE (ed.), *Organic Chemicals from Biomass* (1983); K.L. CHENG, KEIHEI UENO, and TOSHIAKI IMAMURA (eds.), *CRC Handbook of Organic Analytical Reagents* (1982); A. BENNINGHOVEN (ed.), *Ion Formation from Organic Solids: Proceedings of the Second International Conference, Münster, Federal Republic of Germany, September 7–9, 1982* (1983). T.W. GRAHAM SOLOMONS, *Organic Chemistry*, 3rd ed. (1984), is a comprehensive survey.

(C.R.No.)

Hydrocarbons: Brief discussions of the structure, nomenclature, properties, preparation, and reactions of hydrocarbons are presented in all introductory organic chemistry texts. The following are recommended: J.B. HENDRICKSON, D.J. CRAM, and G.S. HAMMOND, *Organic Chemistry*, 3rd ed. (1970); C.R. NOLLER, *Chemistry of Organic Compounds*, 3rd ed. (1965); J.D. ROBERTS and M.C. CASERIO, *Basic Principles of Organic Chemistry* (1964); and L.F. and M.A. FIESER, *Advanced Organic Chemistry* (1961). Special classes of hydrocarbons are discussed in these more advanced treatises: F.D. ROSSINI, B.J. MAIR, and A.J. STREIFF, *Hydrocarbons from Petroleum* (1953), a classic work that describes the isolation of pure hydrocarbons from petroleum; B.T. BROOKS *et al.* (eds.), *The Chemistry of Petroleum Hydrocarbons*, 3 vol. (1954–55); S. PATAI (ed.), *The Chemistry of Alkenes* (1964; H.G. VIEHE (ed.), *Chemistry of Acetylenes* (1969); E.J. CLAR, *Polycyclic Hydrocarbons*, 2 vol. (1964), mainly on compounds with fused six-membered rings (vol. 1 includes a chapter by R. SCHOENTAL on the cancer-producing properties of certain hydrocarbons). Later research is reflected in LOUIS SCHMERLING, *Organic and Petroleum Chemistry for Nonchemists* (1981); K.E. STERIN, V.I. ALEKSANYAN, and G.N. ZHIZHIN, *Raman Spectra of Hydrocarbons* (1980; trans. from the Russian); ALF BJØRSETH (ed.), *Handbook of Polycyclic Aromatic Hydrocarbons* (1983); CLARENCE D. CHENG, *Hydrocarbons from Methanol* (1983); S.O. SCHLANGER and M.B. CITA (eds.), *Nature and Origin of Cretaceous Carbon-Rich Facies* (1982); JOSEPH J. GAJEWSKI, *Hydrocarbon Thermal Isomerizations* (1981).

(H.Ha./Ed.)

Alcohols, phenols, and ethers: UNION CARBIDE CORP., *Alcohols* (1969); R.L. BENT, "Alcohols," in *Encyclopedia of Chemical Technology*, 2nd ed., vol. 1, pp. 531–541 (1963); J.A. MONICK, *Alcohols* (1968); E.H. RODD (ed.), *Chemistry of Carbon Compounds*, 2nd ed., vol. 1A–2D (1965–70).

L.F. and M. FIESER, *Advanced Organic Chemistry*, ch. 22 (1961); F.M. KONRAD, "Phenols," in *Encyclopedia of Polymer Science and Technology*, vol. 10, pp. 73–92 (1969); S.J. LEDERMAN and N. POFFENBERGER, "Phenol," in *Encyclopedia of Chemical Technology*, 2nd ed., vol. 15, pp. 147–159 (1968); E.H. RODD (ed.), *Chemistry of Carbon Compounds*, vol. 3, pt. A, ch. 8 (1954).

UNION CARBIDE CORP., *Ethers* (1960); L.F. and M. FIESER, *Advanced Organic Chemistry*, pp. 303–314 (1961); A.P. LURIE, "Ethers," in *Encyclopedia of Chemical Technology*, 2nd ed., vol. 8, pp. 470–498 (1965). See also TOSHIHIKO HANAI, GUNTER ZWEIG, and JOSEPH SHERMA (eds.), *Phenols and Organic Acids* (1982); H.K. MANGOLD and F. PALTAUF (eds.), *Ether Lipids: Biochemical and Biomedical Aspects* (1983); LON J. MATHIAS and CHARLES E. CARRAHER (eds.), *Crown Ethers and Phase Transfer Catalysis in Polymer Science* (1984).

(J.A.Mo.)

Aldehydes and ketones: C.D. GUTSCHE, *The Chemistry of Carbonyl Compounds* (1967), a short treatise, written for the student in undergraduate organic chemistry, covering the structure, the properties, the mechanism of reactions, and the synthetic utility of carbonyl compounds; C.R. NOLLER, *Chemistry of Organic Compounds*, 3rd ed. (1965), a textbook, designed for a high-level first-year course in organic chemistry, that includes material relating to aldehydes and ketones; R.C. FUSON, *Reactions of Organic Compounds* (1962), a textbook for the advanced student that includes material relating to aldehydes and ketones; E.S. GOULD, *Mechanism and Structure in Organic Chemistry* (1959); J.S. HINE, *Physical Organic Chemistry*, 2nd ed. (1962); J. MARCH, *Advanced Organic Chemistry: Reactions, Mechanisms, and Structure* (1968); and E.E. ROYALS, "The Chemistry of Carbonyl Compounds," in *Advanced Organic Chemistry*, pt. 2 (1954), advanced textbooks including material relating to the mechanisms of the reactions of carbonyl compounds; W.P. Jencks, "Carbonyl- and Acyl-group Reactions," in *Catalysis in Chemistry and Enzymology,* pp. 463–554 (1969), a detailed discussion of the mechanism of various types of carbonyl reactions; S. Patai (ed.), *The Chemistry of the Carbonyl Group* (1966) and *The Chemistry of Carboxylic Acids and Esters* (1969), collections of detailed essays covering various aspects of carbonyl chemistry. Later monographs on the topic include Anthony C. Wilbraham and Michael S. Matta, *Introduction to Organic and Biological Chemistry* (1984); Walter W. Linstromberg and Henry E. Baumgarten, *Organic Chemistry,* 5th ed. (1983); Henry Weiner and Bendicht Wermuth (eds), *Enzymology of Carbonyl Metabolism* (1982); Junji Furukawa and Takeo Saegusa, *Polymerization of Aldehydes and Oxydes* (1963); Mackenzie Walser and John R. Williamson (eds.), *Metabolism and Clinical Implications of Branched Chain Amino and Ketoacids: Proceedings of a Symposium* (1981). (C.D.G.)

Carboxylic acids: Several monographs devoted exclusively to carboxylic acids and their derivatives are: S. Patai (ed.), *The Chemistry of Carboxylic Acids and Esters* (1969); and F.D. Gunstone, *An Introduction to the Chemistry and Biochemistry of Fatty Acids and Their Glycerides,* 2nd ed. (1968). Additional material may be found in selected chapters in the following review series: R.T. Holman (ed.), *Progress in the Chemistry of Fats and Other Lipids;* R. Paoletti and D. Kritchevsky (eds.), *Advances in Lipid Research;* and F.D. Gunstone (ed.), *Topics in Lipid Chemistry.* Specific research is reported in Frank D. Gunstone and Frank A. Norris, *Lipids in Foods: Chemistry, Biochemistry, and Technology* (1983); Michael J. Gibney and David Kritchevsky (eds.), *Animal and Vegetable Proteins in Lipid Metabolism and Atherosclerosis* (1983); T.E. Mittler and R.H. Dadd (eds.), *Metabolic Aspects of Lipid Nutrition in Insects* (1983); William W. Christie, *Lipid Analysis: Isolation, Separation, Identification, and Structural Analysis of Lipids,* 2nd ed. (1982). (F.D.G.)

Organic halogen compounds: W.K.R. Musgrave, "Halogen Derivatives of the Aliphatic Hydrocarbons," in S. Coffey (ed.), *Rodd's Chemistry of Carbon Compounds,* 2nd rev. ed., vol. 1A, ch. 3 (1964); W.J. Feast, "Halogen Derivatives of the Aliphatic Hydrocarbons," *ibid.,* vol. 1A, ch. 3 suppl. (1971); W.J. Feast and W.K.R. Musgrave, "Halogen Derivatives of Benzene and Its Homologues," *ibid.,* vol. 3A, ch. 3 (1971)—together these chapters describe the preparation and reactions of halogenated aliphatic and benzenoid hydrocarbons, including some theoretical interpretation of the chemistry involved. W.A. Sheppard and C.M. Sharts, *Organic Fluorine Chemistry* (1969), describes the chemistry of aliphatic, aromatic, and heterocyclic fluorine-containing compounds and their commercial application; R.O.C. Norman and R. Taylor, *Electrophilic Substitution in Benzenoid Compounds* (1965), and L.M. Stock, *Aromatic Substitution Reactions* (1968), describe in detail theories concerning nucleophilic and electrophilic substitution in aromatic compounds. Environmental aspects are discussed in F. Moriarty (ed.), *Organochlorine Insecticides: Persistent Organic Pollutants* (1975); Milos Hudlicky, *Chemistry of Organic Fluorine Compounds,* 2nd rev. ed. (1976); Peter B.D. de la Mare, *Electrophilic Halogenation* (1976); National Research Council. Committee on Impacts of Stratospheric Change, *Halocarbons: Environmental Effects of Chlorofluoromethane Release* (1976); T.M. Sugden and T.F. West (eds.), *Chlorofluorocarbons in the Environment* (1978); William J. Nicholson and John A. Moore (eds.), *Health Effects of Halogenated Aromatic Hydrocarbons* (1979); M.A.G. Kahn and R.H. Stanton (eds.), *Toxicology of Halogenated Hydrocarbons* (1981). (W.K.R.M.)

Organic nitrogen compounds: N.V. Sidgwick, *The Organic Chemistry of Nitrogen,* 3rd ed. rev. by I.T. Miller and H.D. Springall (1966), a comprehensive account of the subject for readers who have at least the equivalent of college-level organic chemistry; P.A.S. Smith, *The Chemistry of Open-Chain Organic Nitrogen Compounds,* 2 vol. (1965–66), more detailed but narrower coverage of the subject and with more references; C.A. Streuli and P.R. Averell, *The Analytical Chemistry of Nitrogen and Compounds,* 2 vol. (1970), comprehensive treatment of this subject; *Kirk-Othmer Encyclopedia of Chemical Technology,* 2nd ed., 22 vol. (1963–70), presents industrial and technological aspects of nitrogen compounds under many individual headings throughout the various volumes; S. Patai, *The Chemistry of the Amino Group* (1968); J.Z. Zabicky (ed.), *The Chemistry of Amides* (1970); Z. Rappoport, *The Chemistry of the Cyano Group* (1971); and H. Feuer (ed.), *The Chemistry of the Nitro and Nitroso Groups,* 2 vol. (1969–70), highly detailed presentations at an advanced level. See also Melvin J. Astle, *Industrial Organic Nitrogen Compounds* (1961); Peter A.S. Smith, *Derivatives of Hydrazine and Other Hydronitrogens Having N—N Bonds* (1983); Frederich R. Benson, *The High Nitrogen Compounds* (1984). (P.A.S.S.)

Organic phosphorus compounds: R.F. Hudson, *Structure and Mechanism in Organophosphorus Chemistry* (1965); A.J. Kirby and S.G. Warren, *The Organic Chemistry of Phosphorus* (1967), modern mechanistic treatments of the subject, far from comprehensive in their compound coverage; G.M. Kosolapoff, *Organophosphorus Compounds* (1950), deals almost exclusively with compound synthesis, now out of date but the only work of its kind; K. Sasse, *Organische Phosphorverbindungen,* vol. 12/1 and 12/2 of E. Mueller (gen. ed.), *Methoden der organischen Chemie* (1963–64), a comprehensive and indispensable source book for the preparation and properties of organic phosphorus compounds. C. Fest and K.-J. Schmidt, *The Chemistry of Organophosphorous Pesticides,* 2nd rev. ed. (1982), is an informative treatment. (A.J.K.)

Organic sulfur compounds: Ralph Connor, "Organic Sulfur Compounds," in Henry Gilman (ed.), *Organic Chemistry: An Advanced Treatise,* 2nd ed., vol. 1 (1943), is comprehensive but now somewhat dated. W.A. Pryor, *Mechanisms of Sulfur Reactions* (1962), focusses on reactions of elemental sulfur but contains valuable sections on its organic compounds. Norman Kharasch (ed.), *Organic Sulfur Compounds,* vol. 1 (1961) and Norman Kharasch and C.Y. Meyers (eds.), *The Chemistry of Organic Sulfur Compounds,* vol. 2 (1966), are collections of chapters on specialized topics. Frederick Challenger, *Aspects of the Organic Chemistry of Sulphur* (1959), emphasizes natural products. H. Goldwhite treats aliphatic compounds in S. Coffey (ed.), *Rodd's Chemistry of Carbon Compounds,* 2nd ed., vol. 1, pt. B, ch. 5 (1965); and A.R. Forrester and J.L. Wardell and D.R. Hogg, treat aromatic compounds in the same work, vol. 3, pt. A, ch. 5–6 (1971). Individual groups of organic sulfur compounds are the subject of articles in the 2nd ed. of the *Kirk-Othmer Encyclopedia of Chemical Technology:* M.B. Berenbaum and R.N. Johnson, "Polymers Containing Sulfur," vol. 16, pp. 253–281 (1968); Julius Fuchs, "Sulfuric and Sulfurous Esters," vol. 19, pp. 483–498 (1969); W.S. MacGregor, "Sulfoxides," vol. 19, pp. 320–337; E.E. Gilbert, "Sulfonic Acids," vol. 19, pp. 311–319; Leonard Doub, "Sulfonamides," vol. 19, pp. 255–279; S.D. Turk, "Thiols," vol. 20, pp. 205–218, 1969). Specialized research is discussed in Barry M. Trost and Lawrence S. Melvin, *Sulfur Ylides: Emerging Synthetic Intermediates* (1975); and Eric Block, *Reactions of Organosulfur Compounds* (1978). (J.V.K.)

Heterocyclic compounds: Modern textbooks of heterocyclic chemistry include: A.R. Katritzky and J.M. Lagowski, *The Principles of Heterocyclic Chemistry* (1967), also available in French, German, Spanish, Italian, Japanese, Russian, and Polish translation, which offers a concise comparative account; R.M. Acheson, *An Introduction to the Chemistry of Heterocyclic Compounds,* 2nd ed. (1967), which places more emphasis on the division of the subject into compound classes; and A. Albert, *Heterocyclic Chemistry,* 2nd ed. (1968), a more detailed treatment. Of the numerous more specialized works available, K. Schofield, *Hetero-Aromatic Nitrogen Compounds: Pyrroles and Pyridines* (1967), treats the fundamental chemistry of two parent classes of compounds; *Advances in Heterocyclic Chemistry,* 16 vol., ed. by A.R. Katritzky and A.J. Boulton (1963–73); and the *Chemistry of Heterocyclic Compounds,* ed. by A. Weissberger and E.C. Taylor (1950–73), are continuing series of reviews. Later monographs include George R. Newkome and William W. Paudler, *Contemporary Heterocyclic Chemistry: Syntheses, Reactions, and Applications* (1982); Alfred Hassner (ed.), *Small Ring Heterocycles* (1983); H.C. van der Plas, *Ring Transformations of Heterocycles,* 2 vol. (1973). (A.R.K.)

Dyestuffs and pigments: Society of Dyers and Colourists, *The Colour Index,* 3rd ed., 4 vol. (1971). F.W. Gibbs, "Invention in Chemical Industries," in Charles Singer *et al.* (eds.), *A History of Technology,* vol. 3, ch. 25 (1957); E.J. Holmyard, "Dyestuffs in the Nineteenth Century," *ibid.,* vol. 5, ch. 12 (1958). J.C. Cain and J.F. Thorpe, *The Synthetic Dyestuffs,* 7th ed. by J.F. Thorpe and R.P. Linstead (1933); H.E. Fierz-David and L. Blangey, *Fundamental Processes of Dye Chemistry* (1949, Eng. trans. from the 5th Austrian ed. of 1943), includes laboratory experiments; H.A. Lubs (ed.), *The Chemistry of Synthetic Dyes and Pigments* (1971; reprint of 1955 ed.); K. Venkataraman, *The Chemistry of Synthetic Dyes,* 8 vol. (1963–78); E.N. Abrahart, *Dyes and Their Intermediates,* rev. ed. (1977); R.L.M. Allen, *Colour Chemistry* (1971); David Patterson (ed.), *Pigments: An Introduction to Their Physical Chemistry* (1967); R.M. Johnston and M. Saltzman, *Industrial Color Technology* (1972). Articles on special topics (*e.g.,* wood dyes, acetate silk dyes, etc.) may be found in the *Kirk-Othmer Encyclopedia of Chemical Technology,* 3rd ed. (1978–), in progress. Ellis Clayton, *Identification of Dyes on Textile Fibres and Detection of Metals in Fibrous Materials, Dyes and Organic Pigments,* rev. ed. (1963); P.A. Koch, *Rezeptbuch für Faserstoff-Laboratorien* (1960; Eng. trans., *Microscopic and Chemical Testing of Textiles: A Practical Manual,* 1963); Walter Garner, *Textile Laboratory Manual,* 3rd ed., 6 vol. (1966–67); Hans Meier, *Die Photochemie der organischen Farbstoffe*

(1963); FRANCES M. HAMER, *The Cyanine Dyes and Related Compounds* (1964); G. BRITTON, *The Biochemistry of Natural Pigments* (1983); D.H. SOLOMON and D.G. HAWTHORNE, *Chemistry of Pigments and Fillers* (1983). (C.H.Gi./Ed.)

Alkaloids: R.H.F. MANSKE and H.L. HOLMES (eds.), *The Alkaloids: Chemistry and Physiology,* 14 vol. (1950–73), a comprehensive series by authorities in the many fields of alkaloid chemistry, updated at intervals; T. KAMETANI, *The Isoquinoline Alkaloids* (1968), a study of 23 classes of isoquinoline alkaloids with physical properties, structures, and references to discovery and syntheses in tabular form; T. SWAIN (ed.), *Chemical Plant Taxonomy,* ch. 14 (1963), taxonomy from the point of view of molecular relations; R.E. ALSTON and B.L. TURNER, *Biochemical Systematics* (1963), chemical taxonomy from the botanical point of view; R. GOUTAREL, *Les alcaloïdes stéroïdiques des apocynacées* (1964), on isolation, structure, and synthesis; I.D. SPENSER, "The Biosynthesis of Alkaloids and Other Nitrogenous Secondary Metabolites," in M. FLORKIN and E.H. STOTZ (eds.), *Comprehensive Biochemistry,* vol. 20 (1968). See also MAURICE SHAMMA, *The Isoquinoline Alkaloids: Chemistry and Pharmacology* (1972); JOHN S. GLASBY, *Encyclopedia of the Alkaloids,* 4 vol. (1975–83); TREVOR ROBINSON, *The Biochemistry of Alkaloids,* 2nd ed. (1981); S. WILLIAM PELLETIER (ed.), *Alkaloids: Chemical and Biological Perspectives* (1983); A. BAERHEIM SVENDSEN and R. VERPOORTE, *Chromatography of Alkaloids* (1983). (R.H.M.)

Isoprenoids: ERNEST GUENTHER, *The Essential Oils,* 6 vol. (1948–52), a compendium for the general reader of the history, sources, isolation, properties, uses, and analyses of the more volatile isoprenoids; SIR JOHN SIMONSEN *et al., The Terpenes,* 5 vol. (1947–57), an authoritative work on the properties, chemistry, and structures of isoprenoids; WILLIAM TEMPLETON, *An Introduction to the Chemistry of the Terpenoids and Steroids* (1969), a text that requires some knowledge of organic chemistry; L. RUZICKA, "History of the Isoprene Rule," *Proc. Chem. Soc.,* pp. 341–360 (1959), a historical account of the application of the isoprene rule to the structures of terpenoids; J.W. CORNFORTH, "Terpenoid Biosynthesis," *Chem. Brit.* 4:102–106 (1968), a good summary for the nonchemist; JOHN H. RICHARDS and JAMES B. HENDRICKSON, *The Biosynthesis of Steroids, Terpenes and Acetogenins* (1964), details of the biosynthesis of isoprenoids (for the chemist). An interested reader will find the following informative: JOHN S. GLASBY, *Encyclopedia of the Terpenoids,* 2 vol. (1982); FROST & SULLIVAN, *Flavor and Fragrance Market in the U.S.* (1982); UNITED NATIONS INDUSTRIAL DEVELOPMENT ORGANIZATION, VIENNA, *Information Sources on Essential Oils* (1979). (R.H.E.)

Steroids: The chemistry of steroids is comprehensively treated in L.F. and M. FIESER, *Steroids* (1959); more recent information appears in several chapters in S. COFFEY (ed.), *Rodd's Chemistry of Carbon Compounds,* 2nd ed., vol. 2, *Steroids,* 2 pt. (1970–71); for biochemical aspects, see M. FLORKIN and E.H. STOTZ (eds.), *Sterols, Bile Acids and Steroids* (1963); or D.M. GREENBERG, *Metabolic Pathways,* vol. 2 (1968); for comprehensive coverage of the biology and chemistry of the Pill, see D. LEDNICER (ed.), *Contraception: The Chemical Control of Fertility* (1969). For contemporary research see AFANASII A. AKHREM and YURII A. TITOV, *Total Steroid Synthesis* (1970; trans. from the Russian); ROBERT T. BLICKENSTAFF, ANIL C. GHOSH, and GORDON C. WOLF, *Total Synthesis of Steroids* (1974); S. GOROG, *Quantitative Analysis of Steroids* (1983). (R.B.C.)

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Chemical Elements

Chemical elements are the fundamental materials of which all matter is composed. From the modern viewpoint, a substance that cannot be decomposed into simpler substances by ordinary chemical processes is, by definition, an element.

This article considers the origin of the elements and their abundances throughout the universe. The geochemical distribution of these elementary substances in the Earth's crust and interior is treated in some detail, as is their occurrence in the hydrosphere and atmosphere. The article also discusses the periodic law and the tabular arrangement of the elements based on it. The various element groups of the periodic table are described with due attention given to the chief properties and applications of their members. For specific information about the compounds of the elements, see the articles CHEMICAL COMPOUNDS, MOLECULES, and CHEMICAL REACTIONS. Details concerning the structure and properties of matter are provided in ATOMS and MATTER. (Ed.)

This article is divided into the following sections:

General observations. At present there are 110 known chemical elements. (The discovery of element 110, reported in 1987 by Soviet scientists, remains unconfirmed, however, and is still considered extremely tentative.) Elements can combine with one another to form a wide variety of more complex substances called compounds. The number of possible compounds is almost infinite; perhaps a million are known, and more are being discovered every day. When two or more elements combine to form a compound, they lose their separate identities, and the product has characteristics quite different from those of the constituent elements. The gaseous elements hydrogen and oxygen, for example, with quite different properties, can combine to form the compound water, which has altogether different properties from either oxygen or hydrogen. Water clearly is not an element because it consists of, and actually can be decomposed chemically into, the two substances hydrogen and oxygen; these two substances, however, are elements because they cannot be decomposed into simpler substances by any known chemical process. Most samples of naturally occurring matter are physical mixtures of compounds. Seawater, for

Elements, compounds, and mixtures

example, is a mixture of water and a large number of other compounds, the most common of which is sodium chloride, or table salt. Mixtures differ from compounds in that they can be separated into their component parts by physical processes; for example, the simple process of evaporation separates water from the other compounds in seawater.

Historical development of the concept of element. The modern concept of an element is unambiguous, depending as it does on the use of chemical and physical processes as a means of discriminating elements from compounds and mixtures. The existence of fundamental substances from which all matter is made, however, has been the basis of much theoretical speculation since the dawn of history. The ancient Greek philosophers Thales, Anaximenes, and Heracleitus each suggested that all matter is composed of one essential principle—or element. Thales believed this element to be water; Anaximenes suggested air; and Heracleitus, fire. Another Greek philosopher, Empedocles, expressed a different belief—that all substances are composed of four elements: air, earth, fire, and water. Aristotle agreed and emphasized that these four elements are bearers of fundamental properties, dryness and heat being associated with fire, heat and moisture with air, moisture and cold with water, and cold and dryness with earth. In the thinking of these philosophers all other substances were supposed to be combinations of the four elements, and the properties of substances were thought to reflect their elemental compositions. Thus, Greek thought encompassed the idea that all matter could be understood in terms of elemental qualities; in this sense, the elements themselves were thought of as nonmaterial. The Greek concept of an element, which was accepted for nearly 2,000 years, contained only one aspect of the modern definition—namely, that elements have characteristic properties.

In the latter part of the Middle Ages, as alchemists became more sophisticated in their knowledge of chemical processes, the Greek concepts of the composition of matter became less satisfactory. Additional elemental qualities were introduced to accommodate newly discovered chemical transformations. Thus, sulfur came to represent the quality of combustibility, mercury that of volatility or fluidity, and salt that of fixity in fire (or incombustibility). These three alchemical elements, or principles, also represented abstractions of properties reflecting the nature of matter, not physical substances.

The important difference between a mixture and a chemical compound eventually was understood, and in 1661 the English chemist Robert Boyle recognized the fundamental nature of a chemical element. He argued that the four Greek elements could not be the real chemical elements because they cannot combine to form other substances nor can they be extracted from other substances. Boyle stressed the physical nature of elements and related them to the compounds they formed in the modern operational way.

Lavoisier's list of elements

In 1789 the French chemist Antoine-Laurent Lavoisier published what might be considered the first list of elemental substances based on Boyle's definition. Lavoisier's list of elements was established on the basis of a careful, quantitative study of decomposition and recombination reactions. Because he could not devise experiments to decompose certain substances, or to form them from known elements, Lavoisier included in his list of elements such substances as lime, alumina, and silica, which now are known to be very stable compounds. That Lavoisier still retained a measure of influence from the ancient Greek concept of the elements is indicated by his inclusion of light and heat (caloric) among the elements.

Seven substances recognized today as elements (*i.e.,* gold, silver, copper, iron, lead, tin, and mercury) were known to the ancients because they occur in nature in relatively pure form. They are mentioned in the Bible and in an early Hindu medical treatise, the *Caraka-saṃhitā*. Sixteen other elements were discovered in the second half of the 18th century, when methods of separating elements from their compounds became better understood. Eighty-two more followed, after the introduction of quantitative analytical methods.

The atomic nature of the elements. Paralleling the development of the concept of elements was an understanding of the nature of matter. At various times in history, matter has been considered to be either continuous or discontinuous. Continuous matter is postulated to be homogeneous and divisible without limit, each part exhibiting identical properties regardless of size. This was essentially the point of view taken by Aristotle when he associated his elemental qualities with continuous matter. Discontinuous matter, on the other hand, is conceived of as particulate—that is, divisible only up to a point, the point at which certain basic units called atoms are reached. According to this concept, also known as the atomic hypothesis, subdivision of the basic unit (atom) could give rise only to particles with profoundly different properties. Atoms, then, would be the ultimate carriers of the properties associated with bulk matter.

The atomic hypothesis

The atomic hypothesis is usually credited to the Greek philosopher Democritus, who considered all matter to be composed of atoms of the four elements—earth, air, fire, and water. But Aristotle's concept of continuous matter generally prevailed and influenced thought until experimental findings in the 16th century forced a return to the atomic theory. Two types of experimental evidence gave support to the atomic hypothesis: first, the detailed behaviour of gaseous substances and, second, the quantitative weight relationships observed with a variety of chemical reactions. The English chemist John Dalton was the first to explain the empirically derived laws of chemical combination by postulating the existence of atoms with unique sets of properties. At the time, chemical combining power (valence) and relative atomic weights were the properties of most interest. Subsequently numerous independent experimental verifications of the atomic hypothesis were carried out, and today it is universally accepted. Indeed, in 1969 individual uranium and thorium atoms were actually observed by means of an electron microscope.

The structure of atoms. Atoms of elemental substances are themselves complex structures composed of more fundamental particles called protons, neutrons, and electrons. Experimental evidence indicates that, within an atom, a small nucleus, which generally contains both protons and neutrons, is surrounded by a swarm, or cloud, of electrons. The fundamental properties of these subatomic particles are their weight and electrical charge. Whereas protons carry a positive charge and electrons a negative one, neutrons are electrically neutral. The diameter of an atom (about 10^{-8} centimetre) is 10,000 times larger than that of its nucleus. Neutrons and protons, which are collectively called nucleons, have relative weights of approximately one atomic mass unit, whereas an electron is only about $^1/_{2000}$ as heavy. Because neutrons and protons occur in the nucleus, virtually all of the mass of the atom is concentrated there. The number of protons in the nucleus is equivalent to the atomic number of the element. The total number of protons and neutrons is called the mass number because it equals the relative weight of that atom compared to other atoms. Because the atom itself is electrically neutral, the atomic number represents not only the number of protons, or positive charges, in the nucleus but also the number of electrons, or negative charges, in the extranuclear region of the atom.

Composition of atoms

The chemical characteristics of elements are intimately related to the number and arrangement of electrons in their atoms. Thus, elements are completely distinguishable from each other by their atomic numbers. The realization that such is the case leads to another definition of an element, namely, a substance, all atoms of which have the same atomic number.

The existence of isotopes. Careful experimental examination of naturally occurring samples of many pure elements shows that not all the atoms present have the same atomic weight, even though they all have the same atomic number. Such a situation can occur only if the atoms have different numbers of neutrons in their nuclei. Such groups of atoms—with the same atomic number, but with different relative weights—are called isotopes. The number of isotopic forms that a naturally occurring element possesses ranges from one (*e.g.*, fluorine) to as many as ten (*e.g.*, tin); most of the elements have at least two isotopes. The atomic weight of an element is usually determined on large numbers of atoms containing the natural distribution of isotopes, and, therefore, it represents the average isotopic weight of the atoms constituting the sample. More recently, precision mass-spectrometric methods have been used to determine the distribution and weights of isotopes in various naturally occurring samples of elements.

(J.J.L.)

ORIGIN OF THE ELEMENTS

Nuclear fusion

The fundamental reaction that produces the huge amounts of energy radiated by the Sun and most other stars is the fusion of the lightest element, hydrogen, its nucleus having a single proton, into helium, the second lightest and second most abundant, with a nucleus consisting of two protons and two neutrons. In many stars the production of helium is followed by the fusion of helium into heavier elements, up to iron. The still heavier elements cannot be made in energy-releasing fusion reactions; an input of energy is required to produce them (see below *Processes producing heavier elements*).

The proportion of different elements within a star—*i.e.*, its chemical composition—is gradually changed by nuclear fusion reactions. This change is initially concentrated in the central regions of the star where it cannot be directly observed, but it alters some observable properties of the star, such as brightness and surface temperature, and these alterations are taken as evidence of what is going on in the interior. Some stars become unstable and discharge some transmuted matter into interstellar space; this leads to a change in the chemical composition of the interstellar medium and of any stars subsequently formed. The main problem concerned with the origin of the chemical elements is to decide to what extent the chemical composition of the stars seen today differs from the initial chemical composition of the universe and to determine where the change in chemical composition has been produced. Reference is made below to the chemical composition of the universe, but most of the observations refer to our own and neighbouring galaxies.

Cosmic abundances of the elements

The relative numbers of atoms of the various elements are usually described as the abundances of the elements. The chief sources of data from which information is gained about present-day abundances of the elements are observations of the chemical composition of stars and gas clouds in the Galaxy, which contains the solar system and part of which is visible to the naked eye as the Milky Way; of neighbouring galaxies; of the Earth, Moon, and meteorites; and of the cosmic rays.

STARS AND GAS CLOUDS

Spectroscopic analysis

Atoms absorb and emit light, and the atoms of each element do so at specific and characteristic wavelengths; a spectroscope spreads out these wavelengths of light from any source into a spectrum of bright-coloured lines, a different pattern identifying each element; when light from an unknown source is analyzed in a spectroscope, the different patterns of bright lines in the spectrum reveal which elements emitted the light. Such a pattern is called an emission, or bright-line, spectrum. When light passes through a gas or cloud at a lower temperature than the light source, the gas absorbs at its identifying wavelengths and a dark-line, or absorption, spectrum will be formed.

Thus, absorption and emission lines in the spectrum of light from stars yield information concerning the chemical composition of the source of light and of the chemical composition of clouds through which the light has travelled. The absorption lines may be formed either by interstellar clouds or by the cool outer layers of the stars. The chemical composition of a star is obtained by a study of absorption lines formed in its atmosphere.

The presence of an element can, therefore, be detected easily, but it is more difficult to determine how much of it there is. The intensity of an absorption line depends not only on the total number of atoms of the element in the atmosphere of the star but also on the number of these atoms that are in a state capable of absorbing radiation of the relevant wavelength and the probability of absorption occurring. The absorption probability can, in principle, be measured in the laboratory, but the whole physical structure of the atmosphere must be calculated to determine the number of absorbing atoms. Naturally, it is easier to study the chemical composition of the Sun than of other stars, but, even for the Sun, after many decades of study, there are still significant uncertainties of chemical composition. The spectra of stars differ considerably, and originally it was believed that this indicated a wide variety of chemical composition. Subsequently, it was realized that it is the surface temperature of a star that largely determines which spectral lines are excited and that most stars have similar chemical compositions.

There are, however, differences in chemical composition among stars, and these differences are important in a study of the origin of the elements. Studies of the processes that operate during stellar evolution enable estimates to be made of the ages of stars. There is, for example, a clear tendency for very old stars to have smaller quantities of elements heavier than helium than do younger stars. This suggests that the Galaxy originally contained little of the so-called heavy elements (elements beyond helium in the periodic table); and the variation of chemical composition with age suggests that heavy elements must have been produced more rapidly in the Galaxy's early history than now. Observations are also beginning to indicate that chemical composition is dependent on position in the Galaxy as well as age, with a higher heavy-element content near the galactic centre.

In addition to stars, the Galaxy contains interstellar gas and dust. Some of the gas is very cold, but some forms hot clouds, the gaseous nebulae, the chemical composition of which can be studied in some detail. The chemical composition of the gas seems to resemble that of young stars. This is in agreement with the theory that young stars are formed from the interstellar gas.

COSMIC RAYS

High energy electrons and atomic nuclei known as cosmic rays reach the Earth from all directions in the Galaxy (see COSMOS: *Cosmic rays*). Their chemical composition can be observed only to a limited extent, but this can give some information about their place of origin and possibly about the origin of the chemical elements.

The cosmic rays are observed to be proportionately richer in heavy elements than are the stars, and they also contain more of the light elements lithium, beryllium, and boron, which are very rare in stars. One particularly interesting suggestion is that transuranium nuclei may have been detected in the cosmic rays. Uranium is element 92, the most massive naturally occurring on Earth; 13 elements beyond uranium (called the transuranium series) have been created artificially. All transuranium nuclei are highly unstable, which would seem to indicate that the cosmic rays must have been produced in the not too distant past.

(R.J.T.)

SOLAR SYSTEM

Direct observations of chemical composition can be made for the Earth, the Moon, and meteorites, although there are some problems of interpretation. The chemical composition of the Earth's crust, oceans, and atmosphere can be studied, but this is only a minute fraction of the mass of the Earth, and there are many composition differences even within this small sample. Some information about the chemical properties of the Earth's unobserved interior can be obtained by the study of the motion of earthquake

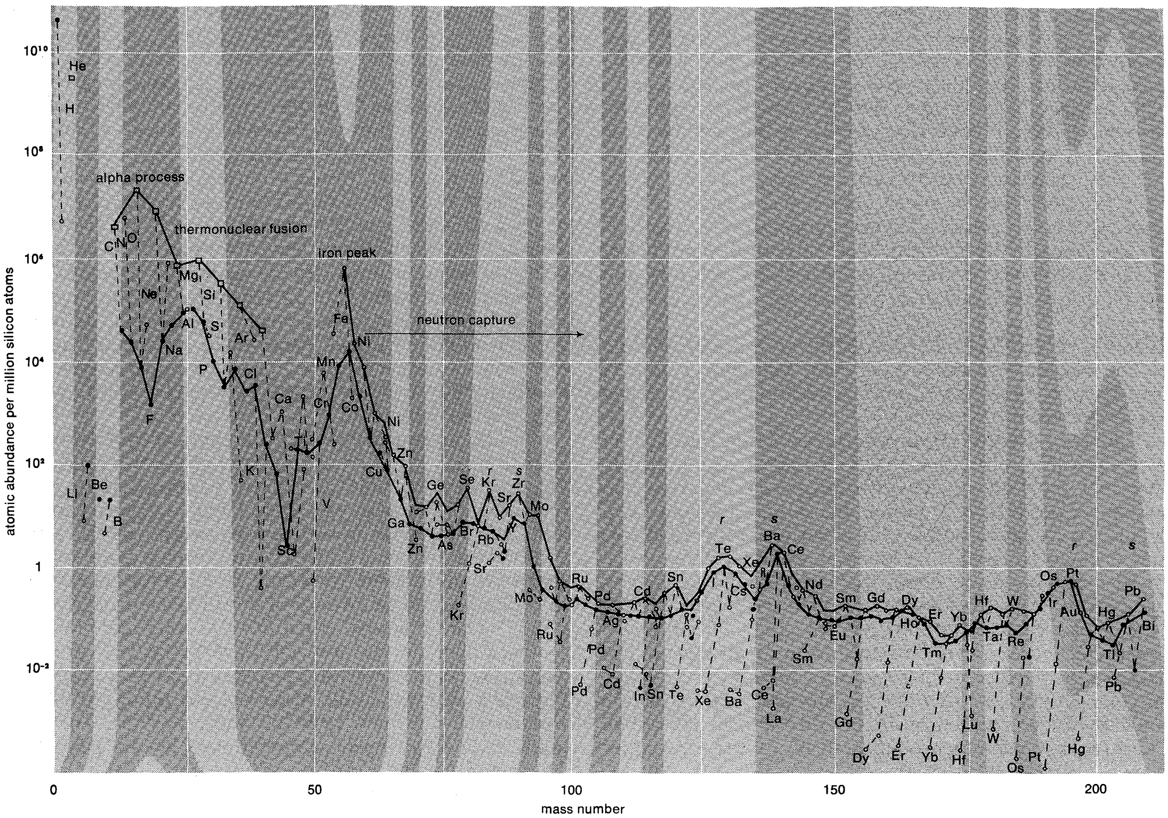

Figure 1: Nuclear abundances plotted against atomic mass number. Filled circles represent nuclei with odd mass numbers; open circles, even mass numbers; open squares, nuclei that can be reduced to helium nuclei (alpha particles). Broken lines join isotopes of given elements; upper solid line at left joins alpha-particle nuclei; upper solid line at right joins nuclei with even mass numbers; the lower solid line joins those with odd mass numbers. The letters *r*, *s* refer to elements formed by the rapid and slow neutron-capture processes.

This diagram first appeared in *New Scientist*, vol. 26, 1965, B. Pagel

waves and by the Earth's magnetic field, which originates in the interior (see below *Geochemical distribution of the elements*).

The Apollo data

Until recently, more was known about element abundances in distant stars than in the Earth's nearest neighbour, the Moon. The lunar landings have provided samples that have been intensively analyzed in many laboratories throughout the world. The data for the Apollo 11 material, collected in the Sea of Tranquility (Mare Tranquillitatis), are given in Table 3. Analyses of Apollo 12 collections are similar for most of the elements. Comparison of the analytical data with those for carbonaceous chondrites (a type of meteorite that provides a good average sample of nonvolatile solar system material) shows that the lunar material has undergone marked geochemical fractionation (segregation of elements). Meteorites suffer from heating in the Earth's atmosphere, so that what is found on Earth is not necessarily the original chemical composition of the meteorites, especially for the volatiles, light gases that are easily lost. When allowance is made for the loss of volatile light gases and for effects of chemical separation, it seems quite possible that the overall chemical composition of the Earth, Moon, Sun, and the meteorites is essentially the same and that they have a common origin.

If elemental abundances are the same in the Earth and stars, isotopic abundances are likely to be the same. Theories predict the relative production of the different isotopes, and it is desirable to be able to compare these with observation. The study of terrestrial abundances of radioactive elements yields information about the age of the solar system, which is discussed below. (R.J.T./Ed.)

SUMMARY OF OBSERVATIONS

The Suess–Urey abundance table

The chemical composition of all objects in the universe is not quite the same, and not all elements can be observed in any one object, even if they are present. Nevertheless, the compositions of many objects are sufficiently similar to make it worthwhile to try to construct a typical table of abundances. Such compilations have been made by several authors and the best known, the work of the American physicists Hans Suess and Harold Urey, is shown in Figure 1. Although this dates from 1956, and later compilations differ in some details, its general character is not in dispute.

The main properties shown in the abundance table (Figure 1) are quite clear. Hydrogen and helium are much more common than all of the other elements. There is a gradual decline toward higher atomic number with a great underabundance of lithium, beryllium, and boron. There is a significant peak in the region of iron, the element with the highest fractional binding energy, and the decline continues to higher atomic number with some subsidiary peaks. These peaks are associated with nuclei containing 50, 82, or 126 neutrons; the theory of nuclear structure predicts that these nuclei should be particularly stable, and these numbers are known as "magic" numbers.

Processes producing heavier elements

As mentioned above, energy can be released by either nuclear fusion or fission reactions and there will be a tendency for material to be gradually converted into elements with maximum binding energy. As observations suggest that hydrogen and helium are much more abundant than other elements, and there is an abundance peak near iron, it is generally supposed that heavy elements have been built up from light elements. In addition, some sites in which element transmutations can occur are known; for example, the interiors of stars tend to get hotter as they evolve, and a succession of nuclear reactions provides the energy that they radiate. Whether or not stars are the site of major nucleosynthesis, some nucleosynthesis certainly occurs there.

Atomic nuclei interact through two strong forces. Be-

Conditions for fusion

cause they have positive electric charges, they repel one another, but there is also a very short-range strong nuclear interaction that is attractive. This may cause fusion reactions to occur if the nuclei ever approach close enough for it to be operative. To overcome the electrical repulsion, the particles must be moving rapidly, as they will be if the material is at a high temperature. The overcoming of the electrical repulsion leads to what are known as thermonuclear reactions. Heavy nuclei have higher electric charges than light nuclei, and a higher temperature is required for reactions between them. The rate of thermonuclear reactions depends on density as well as temperature, but the temperature dependence is much more critical.

REACTION STAGES REFLECTING INCREASING TEMPERATURE

If one imagines a mixture of light elements gradually heated up, a succession of nuclear reactions occurs that is described below.

Hydrogen burning. Hydrogen is converted into helium by a succession of nuclear reactions that change four protons into a helium nucleus, two positrons, and two neutrinos. (A positron is a particle like an electron but with a positive charge; a neutrino is a particle with no charge and negligible mass.) Two different reaction chains exist. In the proton–proton chain the helium nucleus is built up directly from protons. In another series of reactions that involve carbon and nitrogen, called the carbon–nitrogen cycle, the nuclei of carbon and nitrogen are used as catalysts to transform hydrogen into helium; protons are successively added to carbon or nitrogen until a helium nucleus can be emitted by them and the original carbon or nitrogen nucleus reproduced. Both of these reactions occur at temperatures of about 10,000,000 to 20,000,000 K (10,000,000 K is approximately 18,000,000° F).

Helium burning. At temperatures of about 100,000,000 to 200,000,000 K (1 to 2×10^8 K), three helium nuclei can fuse to form carbon. This reaction takes place in the following way: two helium nuclei combine to form an unstable isotope of beryllium, which has an extremely short life; rarely, a third helium nucleus can be added to form carbon before the beryllium decays. Subsequently, a fourth helium nucleus may combine with carbon to give oxygen. The relative amounts of carbon and oxygen produced depend on the temperature and density at which helium is burned.

Carbon and oxygen burning. At temperatures between 5×10^8 K and 10^9 K, pairs of carbon and oxygen nuclei can fuse to produce such elements as magnesium, sodium, silicon, and sulfur.

Silicon burning. Further heating of the material leads to a complicated set of nuclear reactions whereby the elements produced in carbon and oxygen burning are gradually converted into the elements of maximum fractional binding energy; *e.g.,* chromium, manganese, iron, cobalt, and nickel. These reactions have collectively been given the name silicon burning because an important part of the process is the breaking down of silicon nuclei into helium nuclei, which are added in turn to other silicon nuclei to produce the elements noted above.

Reversible nuclear reaction equilibrium. Finally, at temperatures around 4×10^9 K, an approximation to nuclear statistical equilibrium may be reached. At this stage, although nuclear reactions continue to occur, each nuclear reaction and its inverse occur equally rapidly, and there is no further overall change of chemical composition. Thus, the gradual production of heavy elements by nuclear fusion reactions is balanced by disintegrations, and the buildup process effectively ceases once the material is predominantly in the form of iron and its neighbouring elements of the periodic table. Indeed, if further heating occurs, a conversion of heavy nuclei to light nuclei follows in much the same way as occurs in the ionization of atoms when they are heated up. The elements heavier than iron cannot be produced by fusion reactions between light elements; an input of energy is required to produce them.

NEUTRON CAPTURE

It is believed that these heavier elements, and some isotopes of lighter elements, have been produced by successive capture of neutrons. Two processes of neutron capture may be distinguished: the *r*-process, rapid neutron capture; and the *s*-process, slow neutron capture. If neutrons are added to a stable nucleus, it is not long before the product nucleus becomes unstable and the neutron is converted into a proton. Outside a nucleus, a neutron decays into a proton and an electron by a process called beta decay (β-decay). Inside a nucleus it can be stable if the nucleus does not contain too many neutrons. In slow neutron capture, neutrons are added at a rate such that whenever an unstable nucleus is formed, it beta-decays before another neutron can be added. If neutrons can be added more rapidly, as in the *r*-process, the unstable nuclei formed cannot decay before additional neutrons are added until a nucleus is eventually produced that will not accept a further neutron. This nucleus, however, will eventually be subject to beta decay, thus permitting further neutron capture.

Rapid and slow processes of neutron capture

It can be imagined that neutron capture could proceed at an arbitrary rate, giving a mixture of the two processes, but, when the possible sites where neutron-capture reactions could take place are considered, it appears that a fairly clean-cut division between the two processes can be made. If the neutron capture occurs during a quiet stage of stellar evolution, there will be ample time for beta decays to occur, and an *s*-process will result. If neutron capture occurs in an explosive situation, the time scale will be so short that the reaction will have to be an *r*-process. The *r*-process produces the most neutron-rich isotopes of the heavy elements, while those isotopes produced by the *s*-process tend to have relatively more protons. The naturally radioactive nuclei are produced by the *r*-process. The neutron-capture processes appear to give a simple explanation of the magic-number abundance peaks mentioned earlier.

Two small groups of nuclei are not readily fitted into either the sequence of nuclear fusion reactions or the neutron-capture processes. These are nuclei with very low relative abundances. One group consists of the light-nuclei lithium, beryllium, and boron, together with the heavy stable isotope of hydrogen, deuterium. These nuclei are destroyed by nuclear fusion reactions at temperatures lower than that needed to convert hydrogen into helium, and they are bypassed by the production of carbon from helium. The other group consists of the most proton-rich isotopes of some heavy elements, which cannot be produced by the addition of neutrons. Two rather rare or inefficient processes would suffice to produce these isotopes, but there is no complete agreement about what these processes are. It has been suggested that the heavy, proton-rich isotopes might be produced by a process of proton capture and that lithium, beryllium, and boron have been produced by the breakdown of heavier nuclei. A recent suggestion is that they are produced in interstellar space by collisions between cosmic-ray protons and interstellar carbon, nitrogen, and oxygen.

Regions of element synthesis

A discussion of how the present chemical composition of the universe has arisen brings to light two distinct questions: what was the initial chemical composition and what alterations have occurred since creation. Ideally, by working backwards, the initial composition can be deduced from the present composition and a life-history, but this approach is overambitious. The initial composition predicted by simple cosmological theory can then be tested for compatibility with present observations. Element production in the universe as a whole can be discussed first; production in stars and other objects in the Galaxy is treated in the sections that follow.

ELEMENT PRODUCTION IN THE UNIVERSE AS A WHOLE

Hydrogen and helium are overwhelmingly the most abundant elements in the objects of which there is direct knowledge, and, as some buildup of heavy elements occurs in stars, the working hypothesis is usually adopted that the initially created matter contained only light elements. Theories differ only as to whether the initial composition was pure hydrogen or whether there was also a substantial amount of helium.

It is usual to try to devise the simplest theory that can account for the small number of reliable observations of agreed cosmological significance. This need not imply that the universe is simple, but, rather, that it is easiest to work out the consequences of simple theories and there is a better chance of discovering contradictions before too much time has been wasted on detail. The simplest assumption, made in several theories, about the initial chemical composition is that it was pure hydrogen, the hope being that it can be shown that all other elements have been synthesized during the lifetime of the universe.

In the steady-state theory, in which matter is assumed to be continuously created, the initial composition is arbitrary, although supporters of the steady-state theory have always hoped that production of pure hydrogen would give a consistent picture of the universe. In the big-bang theory, which has received much attention in recent years, there is considerably less freedom of choice. Observations of distant galaxies suggest that the universe is expanding and, if this is so, that galaxies may have been very close together about 10,000,000,000 years ago. In the big-bang theory it is assumed that the universe was created at that time and, in the simplest version, that at its creation the universe was very hot as well as very dense. Nuclear reactions in the early stages of the expansion lead to a rather well-defined initial chemical composition for the universe.

Although there are more complicated big-bang theories than the one just mentioned, there is only one arbitrary quantity in the simplest theory, and assignment of some arbitrary specified value to any one of the unknowns will determine the others. If, for example, the density of the universe is specified at some temperature (perhaps 10^{10} K), the subsequent life history of the universe is, in principle, determined. For a wide range of values of this density, the chemical composition after the initial phase is a mixture of hydrogen and helium, with between 20 percent and 30 percent by mass in the form of helium. Most study in recent years has been concerned with trying to decide whether an initial chemical composition devoid of helium or one with 25 percent helium is most likely to be consistent with present observations.

Observed helium content

There are two particular reasons why interest in the big-bang theory has recently been stimulated. The first is concerned with the observed helium content of objects in the Galaxy. It is not always easy to estimate the helium abundance in a star or gas cloud, but most estimates have indicated helium abundances greater than 25 percent by mass. Such values would fit in well with most of the helium being primeval and a small admixture having been produced in stars in the galactic lifetime. The second reason for interest in the big-bang theory is the discovery that very short radio waves, microwaves, are observed to be reaching the Earth from all directions in space. According to the big-bang theory, the universe was filled with radiation in its early stages and most of this radiation has never subsequently been absorbed. As the universe has expanded, the radiation has been shifted toward longer wavelengths by the Doppler effect, a change in wavelength brought about by motion of the source with respect to the observer. As a result of this effect, the radiation created by the big bang would be expected to appear today as microwaves of just the type that have been observed.

The big-bang theory not only predicts that all objects, except those in which the helium can have been destroyed, should have a minimum of about 25 percent helium but that the microwave radiation should have a particular distribution with frequency known as the Planck form. The theory can be proved false either if objects are discovered with a substantially smaller amount of helium than 25 percent or if the microwave spectrum does not have the Planck form. Several observations have been made which suggest that the theory is false, but to date none of them appears to be conclusive.

ELEMENT PRODUCTION IN STARS

A substantial amount of nucleosynthesis must have occurred in stars. It was stated above that a succession of nuclear fusion reactions takes place as the temperature of the stellar material rises. Theories of stellar evolution indicate that the internal temperatures of stars first rise during their life history and eventually fall after reaching a maximum value. For very low-mass stars, the maximum temperature may be too low for any significant nuclear reactions to occur, but for stars as massive as the Sun or greater, most of the sequence of nuclear fusion reactions described above can occur. Moreover, a time scale for stellar evolution is derived in theories of stellar evolution that show that stars substantially more massive than the Sun can have completed their active life history in a time short compared with the age of the universe derived from the big-bang cosmological theory.

This result implies that stars more massive than the Sun, which were formed very early in the life history of the Galaxy, could have produced some of the heavy elements that are seen today but that stars much less massive than the Sun could have played no part in this production. Unless the Galaxy is very much older than is generally believed, such low mass stars, even if formed with the Galaxy, would still be at an early stage in their evolution because changes within them proceed at a relatively slow pace. If there has been substantial heavy-element production in stars, a sufficient fraction of the earliest stars formed must have been relatively massive.

Mass loss in stars

If substantial nucleosynthesis has occurred in stars, could such a process have produced all of the heavy elements that are observed today and possibly all of the helium inside the stars? A vital point is the following: if the heavy elements produced in stars are to influence what is observed, they must be expelled from the interiors of the stars in which they are produced and incorporated into future generations of stars, in which they can be observed subsequently. Unfortunately, direct knowledge of mass loss from stars is fragmentary; steady loss of mass is observed in some stars, and a few are observed to explode catastrophically, as in the explosion of a supernova. At present it is only possible for a very rough estimate to be made of the rate of exchange of matter between stars and the interstellar medium.

Supernovae are believed to be stars reaching the end of their evolution, and many astronomers believe that a supernova explosion is the main process whereby heavy elements produced inside stars are returned to the interstellar medium. In addition, because a supernova explosion is the most violent type of event regularly observed in galaxies, it is believed that cosmic rays must also be produced in the explosion. Some rough estimates follow. The mass of the Galaxy is believed to be between 10^{11} and 2×10^{11} solar masses, and perhaps 2×10^{9} solar masses are heavy elements. If these heavy elements were produced steadily in a galactic lifetime of about 10^{10} years, one-fifth of a solar mass of heavy elements must have been produced each year. Counts of supernovae in nearby galaxies suggest that there might be one supernova explosion per large galaxy about every 30 years. If all the heavy elements are produced in supernovae, about six solar masses are required from each explosion. Although these numbers are very uncertain, this amount seems too large, but it could be reduced if the frequency of supernovae is very much higher in young galaxies. The possibility remains that a significant quantity of heavy elements may be produced by a very large number of less spectacular stars or by much more massive objects that are mentioned below.

If there has been a gradual production of heavy elements, recently formed stars should contain more than old stars. It is possible to identify some stars which have formed quite recently. The light output of stars rises as a rather high power of their mass according to a mass–luminosity relation that is valid for the vast majority of stars whose masses are known, while their supply of nuclear energy is only directly proportional to the mass. This means that the more massive stars complete their life history much more rapidly than low-mass stars and that the brightest stars observed today cannot be more than a few million years old at the most. The heavy-element content of the young stars is greater than that of many old stars, perhaps because of a gradual increase in the heavy-element content of the interstellar medium from which stars are formed. Observations show that only the very oldest stars have an extremely small amount of very heavy elements in their visible layers, and it appears that element production must have been much more rapid when the Galaxy was young than it is now. There may indeed have been a much higher frequency of supernovae. Recent observations suggest also that chemical composition is a function of a star's place of origin as well as its age. In particular, the production of heavy elements may have been higher near the centre of the Galaxy than elsewhere (see below *Element production in massive objects*).

Although the first nuclear reaction to occur in stars is the conversion of hydrogen into helium, all of the helium that is observed today can hardly have been produced in ordinary stars, the more so if all objects contain more than about 25 percent helium by mass. Considering the relative amounts of helium and heavier elements, observations indicate that the total mass of helium may be ten times greater than that of the heavier elements; if all elements other than hydrogen have been produced in stars, the relative production of helium and heavier elements must have just this value. As stars evolve, however, the conversion of hydrogen into helium is followed by the conversion of helium into heavier elements. At all stages in a star's evolution there will be a region where the temperature is suitable for the conversion of hydrogen into helium, but it appears that there will be only a thin shell of helium separating the regions in which hydrogen has not yet been converted into helium and the region where helium has been burned into heavy elements. The possible chemical composition of a highly evolved star is a series of layers of different chemical composition. The central region would contain elements such as iron and nickel with layers of successively lighter elements surrounding it and the outermost layer containing essentially only hydrogen or hydrogen and helium. A very special type of mass loss would be required to expel 10 times as much helium as heavy elements from these different layers into interstellar space.

Layers of elements

It is also difficult to see how the full amount of helium could have been produced. If a quarter of the galactic mass, originally hydrogen, has been converted into helium, it can be shown that essentially all of the mass must have passed through at least one generation of massive stars. The total energy release under such a circumstance would imply that the Galaxy was very much more luminous in the past—one hundred times more luminous for the first 10 percent of its lifetime, for example.

ELEMENT PRODUCTION IN MASSIVE OBJECTS

Collapse in supermassive stars

Although there is no direct evidence for the existence of stars more than about 50 times as massive as the Sun, there is no obvious reason why much more massive objects should not exist. If they were sufficiently massive, they would not behave as ordinary stars because their gravitational attraction would be so strong that not even the energy released by conversion of hydrogen into helium would prevent such supermassive stars from continuing to collapse rapidly. According to present theoretical ideas, if such a collapse is spherically symmetrical, nothing can prevent the supermassive object from collapsing to an extremely high or infinite density; but, if it is asymmetrical—because it is, for example, rapidly rotating—there is some possibility that the catastrophic collapse, called an implosion, might be followed by explosion. At the high-density, high-temperature phase of such an object, some nucleosynthesis (manufacture of nuclei from smaller nuclei) would occur, primarily of helium but with a small amount of heavier elements according to the arguments given early in this article. Such objects have been suggested as a possible important source of helium.

There is some observational evidence that explosions on a very much greater scale than single supernovae are occurring in galaxies. In some peculiar galaxies that are strong emitters of radio waves, there is evidence that explosions have thrown a large quantity of gas hundreds of thousands of light-years into intergalactic space. Such galactic explosions may not be related to the theoretical supermassive objects mentioned above, but it is difficult to believe that some nucleosynthesis does not take place during the phases of extreme conditions that must occur in such objects. The suggestion that heavy-element abundances may be higher near the centre of the Galaxy could be related to a past explosion there.

Radioactive chronologies

Radioactive elements in the Earth, the Moon, and in meteorites can provide useful information about the ages of these objects and about the dates of formation of the heavy elements themselves. The elements uranium and thorium gradually decay into lead, different isotopes of lead arising from the various isotopes of uranium and thorium; some isotopes of lead are, however, not produced by any radioactive decay process. When the rocks of the Moon or the Earth's crust or the meteorites solidified, further chemical separation of the radioactive elements and their decay products was prevented. By studying the relative amounts of the radioactive isotopes and their decay products, it is possible to obtain an estimate of when the rocks solidified. Estimates can also be made using radioactive isotopes other than uranium and thorium.

The results of these discussions indicate that the meteorites, or at least the parent body of the meteorites, solidified between 4.5×10^9 and 4.6×10^9 years ago. It is possible to speak with such confidence of this age because two isotopes of uranium and one of thorium have very different decay times that bracket that value. There is no unique age for the rocks of the Earth's crust because there has been considerable volcanic activity during the Earth's history and rocks have solidified at all stages. All indications are that the oldest rocks have ages of the same order as the ages of the parent bodies of the meteorites. Only a very small region of the Moon's surface has been studied so far, but it has been found to have very old rocks of age up to about 4.5×10^9 years. No conclusions can be drawn about the date of solidification of the Moon from these few observations, as nothing is known about its past geological history, but they are certainly not inconsistent with the view that the Earth, the Moon, and meteorites have a similar age and origin.

Radioactive nuclei

It has also been found possible to obtain information about the time of formation of the radioactive elements. Assuming that both radioactive nuclei and their stable neighbours are produced by the neutron-capture process discussed earlier, theory predicts a relative production rate for all of the nuclei. The radioactive nuclei can be divided into three groups: short-lived, medium-lived, and long-lived, where short-lived means considerably less than the believed age of the universe and long-lived means comparable with that age. If radioactive nuclei are produced and decay steadily, then at some point in time the total amount of a short-lived isotope reaches a steady value. In meteorites, one can study the decay products of such short-lived nuclei and can discover their abundance when the meteorites were formed. This amount is lower than the expected value, suggesting that nucleosynthesis ceased in the solar system material about 2×10^8 years before the meteorites and planets solidified.

Study of the decay products of nuclei with medium decay rates indicates that their abundance is higher than if nucleosynthesis has occurred at a constant rate throughout galactic history. This suggests that the solar system material was significantly enriched in heavy elements shortly before the cessation of nucleosynthesis—that is, before the Sun and planets were formed. Finally, the very long-lived isotopes give information about the total time scale of nucleosynthesis that is not inconsistent with the galactic age estimated by other methods.

Although there is not unanimous agreement concerning these results, it appears that it is, in principle, possible to obtain a considerable amount of information about the past rate of nucleosynthesis and possibly about the types of objects in which it has occurred. In particular, it may eventually be possible to decide whether most element production has occurred in a large number of supernovae or in a much smaller number of massive objects.

(R.J.T.)

GEOCHEMICAL DISTRIBUTION OF THE ELEMENTS

Basic work of Clarke and Goldschmidt

Knowledge of the geochemical distribution of elements involves elucidation of the relative and absolute abundances of the chemical elements in the Earth and in its various parts—the crust, interior, atmosphere, and hydrosphere. This comprises a major part of the science of geochemistry, which is the study of the distribution of the chemical elements in space and time and the laws governing this distribution. Basic knowledge in this area was largely accumulated during the 19th century. As noted above, the concept of a limited number of chemical elements had been established by 1800, and the appearance of the periodic table, in 1869, provided a new insight into the limitations on the number of elements. Concurrent with these advances in chemical understanding, from about 1850 onward there was a steadily increasing output of analytical data on the Earth's rocks, minerals, and waters, mainly from laboratories in Europe and North America. The output from North America was materially increased following the establishment of the United States Geological Survey in 1879 and the appointment of Frank W. Clarke as chief chemist in 1884.

Clarke's name will always be linked with the study of the geochemical distribution of the elements—indeed, the term clarke was proposed as the unit for the average percentage of an element in the Earth's crust by Soviet scientists and has been generally adopted. In 1889 Clarke wrote the first of his many publications on the geochemical distribution of the elements. He assembled many chemical analyses of rocks from different continents, calculated average values, and showed that the overall chemical compositions of continental areas are remarkably similar. By combining these averages he obtained values for the abundances of the commoner elements in the continental crust of the Earth, values that have not been materially changed in spite of the vast increase of available data since that time. He also estimated abundances for many of the less common elements; these estimates were based in many instances on very limited and imprecise data and subsequently have been improved.

A further development of great significance was the assemblage of comprehensive data on the abundances of individual elements in terrestrial materials and in the Cosmos (based on solar and meteorite abundances) by the Norwegian geochemist Victor Moritz Goldschmidt during the 1930s. Goldschmidt's tables provided the basis for modern research on the geochemical distribution of the elements, and his compilation of data on cosmic abundances was the key to later theories on element synthesis in stars and supernovae.

Goldschmidt also contributed to the understanding of elemental distribution within the Earth through his geochemical classification of the elements into lithophile, siderophile, chalcophile, and atmophile. Lithophile elements are those with a strong affinity for oxygen; they are concentrated in the crust or lithosphere as silicate and oxide minerals. Siderophile elements are principally metals that alloy readily with iron; Goldschmidt explained their scarcity in the Earth's crust by their concentration in the nickel–iron core (the siderosphere). Chalcophile elements are those with strong affinity for sulfur; they occur mainly as sulfides. And atmophile elements are gases, such as nitrogen, argon, and other rare gases, which are unreactive and hence accumulate in the atmosphere. (Goldschmidt also proposed a group of biophile elements, for those that concentrate in living matter—essentially carbon, hydrogen, oxygen, nitrogen, sulfur, and phosphorus.)

Terrestrial distribution

The study of earthquake waves passing through the body of the Earth has shown that the interior is not uniform; it consists of distinct shells separated by concentric discontinuities at which the velocities of the passing waves change. The two major discontinuities that are universally recognized are the Mohorovičić Discontinuity, which divides the Earth's crust from its underlying mantle, and the Wiechert–Gutenberg Discontinuity, which separates the mantle from the core. The latter discontinuity exists at a depth of 2,900 kilometres (1,800 miles); it is marked by a sudden increase in density, from about 5.7 at the base of the mantle to 9.7 at the top of the core. The only reasonable interpretation of this discontinuity is that the mantle consists of silicates and oxides of the common elements (largely magnesium and iron), and the core consists of metallic iron alloyed with minor amounts of other elements (analogous to the nickel-iron in meteorites). The Mohorovičić Discontinuity varies in depth from place to place; it averages about 33 kilometres (20 miles) below the continents and about 8 kilometres (5 miles) below the bottom of the deep oceans. It too is marked by a density increase from crust to mantle—a comparatively small one, from about 3 to 3.3.

To the three spherical divisions—crust, mantle, and core—two more should be added: the hydrosphere, which is the discontinuous shell of fresh and salt water, on and within the crust; and the atmosphere, the ocean of air that surrounds the Earth, gradually thinning into the vacuum of outer space. The principal characteristics of these "geospheres" are set forth in Table 1.

Table 1: Volume and Masses of Earth Shells

	thickness (km)	volume ($\times 10^{27}$ cm^3)	mean density (g/cm^3)	mass ($\times 10^{27}$ g)	mass (percent)
Atmosphere	—	—	—	0.000005	0.00009
Hydrosphere	3.73 (mean)	0.00135	1.03	0.00139	0.024
Crust	20 (mean)	0.010	2.8	0.029	0.5
Mantle	2,880	0.897	4.5	4.010	67.1
Core	3,471	0.175	11.0	1.936	32.4
Whole Earth	6,371	1.083	5.52	5.976	100.00

THE EARTH'S CORE

The evidence for the composition of the core is all indirect because no means have yet been devised for directly sampling the deep interior of the Earth. The moment of inertia of the Earth indicates that there is a concentration of mass around the centre, and seismic data have shown that below the Wiechert–Gutenberg Discontinuity the density of the material is high, ranging upwards from 9.7. The only heavy element with high cosmic abundance is iron, and because an iron–nickel alloy is an important meteorite component, it is reasonable to conclude that the Earth's core consists largely of metallic iron with a minor admixture of other elements. This conclusion is supported by geophysical evidence that indicates that the mean atomic number of the material of the core is about 22. The atomic number of iron is 26, so this implies that the core also contains elements of lower atomic number. Sulfur, with atomic number 16, and carbon, 6, are relatively abundant in meteoritic matter, and the presence of minor amounts of these elements in the core would effectively reduce the mean atomic number. Some authorities have advocated silicon (atomic number 14) as the major alloying component in the core, but this seems less likely; if silicon were the sole alloying element, then the core would have to contain more than 30 percent silicon in order to reduce its mean atomic number to 22. In addition, free silicon requires extremely reducing conditions (lack of oxygen), and the presence of ferrous iron in the mantle is inconsistent with this requirement.

It is not possible to give definite figures for the abundances of the elements in the Earth's core. It is certainly made up largely of metallic iron, however, probably with some nickel, a little cobalt, and appreciable amounts of such lighter elements as carbon and sulfur.

THE EARTH'S MANTLE

The mantle comprises that part of the Earth between the Mohorovičić and the Wiechert–Gutenberg discontinuities.

It makes up 83 percent of the volume of the Earth and 67 percent of its mass and is thus of decisive importance in determining the bulk composition of the planet. In estimating elemental abundances in the mantle, however, the same difficulty as with the core arises: direct sampling is not feasible. Much more geophysical data are available for the mantle, however, and some volcanic eruptions have brought rock fragments to the surface that have certainly been derived from this zone. The most remarkable of these materials are the diamond-bearing inclusions found in the famous pipes, or volcanic necks, that are mined in South Africa and Siberia. The presence of diamond, the high-pressure form of carbon, implies a depth of origin of at least 100 kilometres (62 miles), but these inclusions are rare. The common type of mantle-derived inclusion is peridotite, a silicate rock consisting largely of olivine, $(Mg,Fe)_2SiO_4$, with minor amounts of orthopyroxene, $(Mg,Fe)SiO_3$, and diopside, $CaMg(Si_2O_6)$.

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Geophysical information indicates that below a depth of about 1,000 kilometres (620 miles), the mantle behaves as an essentially homogeneous material, but above this level its physical properties are more varied, and there is evidence for second-order discontinuities. This region above 1,000 kilometres is frequently referred to as the upper mantle, and in recent years has been the object of a concentrated research effort by geologists and geophysicists all over the world. The significance of the upper mantle is that processes originating there have dramatic effects on the surface—in the form of volcanic eruptions and some earthquakes—and less dramatic but equally important effects within the crust, such as the introduction and concentration of some elements, possibly leading to the formation of ore deposits. Increased knowledge of the upper mantle thus has both scientific and economic appeal.

Geophysical data on the properties of the upper mantle suggest that it must consist essentially of magnesium-iron silicates, probably largely olivine in the region immediately below the crust. Olivine is not stable under very high pressures, however; it is converted to a different phase of about 10 percent higher density and with a structure like the mineral oxide spinel ($MgAl_2O_4$). This conversion would occur in the mantle at depths of around 400 kilometres, and a second-order discontinuity at that depth can plausibly be ascribed to this conversion. Pyroxenes also undergo transformations to phases of greater density at the high pressures within the mantle. Thus the mantle, although composed of material of familiar chemical composition, consists, in its lower part at least, of different minerals than those in the upper part.

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Many estimates of the composition of the upper mantle have been made in recent years; some of these are presented in Table 2. On the whole, the similarities are more important than the differences. All agree that the principal components are oxides of silicon, magnesium, and iron. The differences are mainly in the minor components such as aluminum oxide, calcium oxide, and the alkalies, and are determined largely by theoretical considerations and the weight given to specific aspects of the geophysical and geochemical data.

Although fairly reliable estimates exist for the abundances of the major elements in the mantle, little is known of minor and trace elements. Knowledge of the crystal structure of possible mantle minerals indicates that many minor and trace elements will not be readily incorporated, however. They are therefore likely to concentrate in liquid material in the mantle and be carried upward in solution, eventually being transported into the crust. It is thus probable that the mantle is relatively depleted, and the crust relatively enriched, in minor and trace elements. This is certainly true for uranium and thorium, because the amount of these elements in the crust is almost sufficient to account for the total amount of heat flowing out of the Earth.

THE EARTH'S CRUST

The crust is a comparatively thin shell on the surface of the Earth and makes up less than 1 percent of its total mass. Its geochemical significance is only marginally related to its bulk, however. It has been subjected to extensive investigations, and it provides the raw materials on which civilization depends. It is the most diverse of the geospheres, being a complex mosaic of many rock types—igneous, sedimentary, and metamorphic—each with a wide variety of chemical and mineralogical compositions. The surface is veneered with soils, related in composition to the rocks from which they formed, but with important modifications because of the smaller grain size, the presence of organic matter, and an intricate complex of living organisms. Ultimately, man's welfare and indeed his survival depends on the wise utilization of the materials in the crust. Modern civilization has been erected upon the exploitation of fuels and ore deposits, which are simply geochemical concentrations of useful elements.

Igneous rocks. Clarke estimated that 95 percent of crustal rocks are of igneous origin (formed from molten silicate masses, or magmas). Sedimentary rocks occur as a thin veneer on an igneous or metamorphic basement, except where locally thickened in mountain belts. The primordial rocks of the crust must have been essentially igneous, and the first sedimentary rocks were derived from them by processes of weathering and erosion. Metamorphic rocks are formed from both sedimentary and igneous rocks by transformations due to heat and pressure at depth in the crust; unless very intense, these transformations do not totally obliterate the primary igneous or sedimentary features. Igneous rocks show a wide range of composition; the principal component, silica (SiO_2), ranges from about 35 percent to 80 percent among the commoner igneous rocks, and other components also show a wide variation. They thus illustrate some quite extensive geochemical fractionations of the elements, the fractionations that may have economic significance if they bring about the findings of workable ore deposits.

Chief components of the upper mantle

In 1924 a comprehensive review of igneous rock composition based on compilation of over 5,000 superior analyses was published. This was in many respects the ultimate refinement of Clarke's initial review of 1889. It confirmed that the averages of analyses from different

Table 2: Estimated Composition of the Earth's Mantle
(weight percent)

	estimated from peridotite nodules	calculated from three parts peridotite and one part basalt	based on the composition of oceanic basalts	estimated from peridotite nodules	calculated from a meteoritic Earth model
Silicon dioxide (SiO_2)	42.86	45.16	45.1	44.2	48.09
Titanium dioxide (TiO_2)	0.33	0.71	0.5	0.1	0.13
Aluminum oxide (Al_2O_3)	6.99	3.54	4.1	2.7	3.02
Chromic oxide (Cr_2O_3)	0.18	0.43	0.3	0.30	0.55
Ferric oxide (Fe_2O_3)	0.36	0.46	2.0	1.1	—
Ferrous oxide (FeO)	8.97	8.04	7.9	7.3	12.71
Manganese oxide (MnO)	0.14	0.14	0.2	0.15	0.43
Nickel oxide (NiO)	0.20	0.20	0.2	0.20	—
Magnesium oxide (MgO)	35.07	37.47	36.7	41.3	31.15
Calcium oxide (CaO)	4.37	3.08	2.3	2.4	2.32
Sodium oxide (Na_2O)	0.45	0.57	0.6	0.25	1.13
Potassium oxide (K_2O)	0.003	0.13	0.02	0.015	0.13
Phosphate (P_2O_5)	—	0.06	0.1	—	0.34

continental areas are essentially identical. It also revealed significant geochemical differences between the continental and oceanic crusts. The average of igneous rock analyses from the oceanic islands is notably lower in silica and alkalies, and higher in magnesium and calcium oxides, than the continental averages. This is simply a reflection of the fact that most oceanic islands, such as Hawaii, consist almost entirely of basalts (averaging about 50 percent silica), whereas continental areas include large granitic masses, with silica contents around 70 percent. In terms of volumes, igneous rocks consist predominantly of two great types, granitic and basaltic. The former essentially are confined to the continents and the latter occur in both continents and ocean basins. The other types of igneous rocks, while many and varied, are quantitatively insignificant and hardly affect the averages. Thus, for the major elements, the average of over 5,000 analyses of igneous rocks is not significantly different from the simple aver-

Table 3: Abundances of the Elements in Different Geochemical Materials*†
(figures are in parts per million unless otherwise indicated)

atomic number and element	carbonaceous chondrites (type I)	Moon (Apollo 11)	Earth's crust	granite (G-1)	basalt (W-1)	shales	sandstones	carbonates	metamorphic rocks	seawater
(1) Hydrogen	2.2%	200	1,400	380	590	—	—	—	1,000	10.7%
(3) Lithium	1.3	12	20	24	12	66	15	5	22	0.17
(4) Beryllium	0.04	2	2	3	0.8	3	0.X	0.X	1.3	6×10^{-7}
(5) Boron	5	2	10	1.5	15	100	35	20	—	4.5
(6) Carbon	3.7%	140	250	260	220	1.0%	—	11.3%	0.37%	28
(7) Nitrogen	2,700	100	20	8	14	600	—	—	—	0.7
(8) Oxygen	45.3%	40.0%	46.6%	48.2%	44.6%	50.5%	53.4%	49.8%	48.0%	85.7%
(9) Fluorine	190	140	650	700	250	800	180	240	500	1.3
(11) Sodium	0.55%	0.33%	2.8%	2.4%	1.6%	0.96%	0.33%	0.04%	2.6%	1.08%
(12) Magnesium	9.6%	4.5%	2.1%	0.23%	4.0%	1.5%	0.70%	4.7%	1.4%	0.13%
(13) Aluminum	0.85%	5.6%	8.1%	7.4%	7.9%	8.0%	2.5%	0.42%	7.7%	1×10^{-2}
(14) Silicon	10.3%	19.2%	27.7%	34.0%	24.9%	7.3%	36.8%	2.4%	30.4%	2.9
(15) Phosphorus	1,400	500	700	390	610	700	170	400	650	9×10^{-2}
(16) Sulfur	6.2%	1,700	260	170	130	2,400	240	1,200	600	904
(17) Chlorine	260	14	130	70	200	180	10	150	100	1.94%
(19) Potassium	0.05%	0.14%	2.6%	4.6%	0.53%	2.7%	1.1%	0.27%	2.6%	392
(20) Calcium	1.06%	8.0%	3.6%	1.00%	7.8%	2.2%	3.9%	30.2%	2.9%	411
(21) Scandium	5	75	20	3	34	13	1	1	—	4×10^{-5}
(22) Titanium	0.04%	5.9%	0.44%	0.16%	0.64%	0.46%	0.15%	0.04%	0.31%	1×10^{-3}
(23) Vanadium	57	50	140	16	240	130	20	20	53	2×10^{-3}
(24) Chromium	2,200	2,100	100	22	120	90	35	11	99	2×10^{-4}
(25) Manganese	1,700	1,900	950	230	1,300	850	X0	1,100	530	4×10^{-4}
(26) Iron	18.4%	14.3%	5.0%	1.36%	7.7%	4.7%	0.98%	0.38%	3.1%	4×10^{-3}
(27) Cobalt	480	25	25	2.4	50	19	0.3	0.1	21	4×10^{-4}
(28) Nickel	1.04%	15	70	1.5	78	68	2	20	23	7×10^{-3}
(29) Copper	140	11	50	13	110	45	X	4	14	1×10^{-3}
(30) Zinc	320	15	70	45	82	95	16	20	—	5×10^{-3}
(31) Gallium	10	4.5	15	18	16	19	12	4	14	3×10^{-5}
(32) Germanium	34	0.1	1.5	1	1.7	1.6	0.8	0.2	—	6×10^{-5}
(33) Arsenic	2.0	0.05	1.8	0.8	2.4	13	1	1	—	3×10^{-3}
(34) Selenium	27	0.8	0.1	—	0.11	0.6	0.05	0.08	—	1×10^{-4}
(35) Bromine	5	0.1	2.5	0.5	0.4	4	1	6	—	67
(37) Rubidium	2.3	3.4	90	220	22	140	60	3	118	0.12
(38) Strontium	8	170	390	250	180	300	20	610	340	8.1
(39) Yttrium	1.6	120	30	13	25	30	40	30	—	1×10^{-5}
(40) Zirconium	11	370	170	210	100	160	220	19	400	3×10^{-5}
(41) Niobium	0.5	21	20	20	10	11	0.0X	0.3	—	2×10^{-5}
(42) Molybdenum	1.6	0.5	1.5	7	0.5	2.6	0.2	0.4	—	1×10^{-2}
(44) Ruthenium	0.7	—	0.01	—	<0.4	—	—	—	—	7×10^{-7}
(45) Rhodium	0.2	—	0.005	<0.005	<0.005	—	—	—	—	—
(46) Palladium	0.6	0.006	0.01	0.002	0.01	—	—	—	—	—
(47) Silver	0.4	0.008	0.05	0.04	0.05	0.07	0.0X	0.0X	—	3×10^{-4}
(48) Cadmium	1.0	0.004	0.2	0.06	0.3	0.3	0.0X	0.04	—	1×10^{-4}
(49) Indium	0.09	—	0.05	0.025	0.07	0.1	0.0X	0.0X	—	—
(50) Tin	1.6	0.6	3	4	3	6	0.X	0.X	—	8×10^{-4}
(51) Antimony	0.15	0.007	0.2	0.4	1.1	1.5	0.0X	0.2	—	3×10^{-4}
(52) Tellurium	3.3	0.02	0.01	—	<0.2	—	—	—	—	—
(53) Iodine	0.6	—	0.5	—	<0.05	2.2	1.7	1.2	—	6×10^{-2}
(55) Cesium	0.19	0.10	2	1.5	1.0	5	0.X	0.X	—	3×10^{-4}
(56) Barium	4	200	400	1,200	180	580	X0	10	1,070	2×10^{-2}
(57) Lanthanum	0.19	18	30	100	12	34	30	X	—	3×10^{-6}
(58) Cerium	0.63	54	60	170	23	67	92	12	—	1×10^{-6}
(59) Praseodymium	0.09	11	8	17	4	8.8	8.8	1.1	—	6×10^{-7}
(60) Neodymium	0.42	46	30	55	17	31	37	4.7	—	3×10^{-6}
(62) Samarium	0.13	15	6	9	4	6.9	10	1.3	—	1×10^{-7}
(63) Europium	0.05	1.9	1	1.3	1.1	1.4	1.6	0.2	—	1×10^{-7}
(64) Gadolinium	0.24	20	5	5	4	6.2	10	1.3	—	7×10^{-7}
(65) Terbium	0.04	3.6	1	0.6	0.8	1.0	1.6	0.2	—	1×10^{-7}
(66) Dysprosium	0.22	25	3	2.5	4	4.5	7.2	0.9	—	9×10^{-7}
(67) Holmium	0.06	4.9	1	0.5	1	1.2	2.0	0.3	—	2×10^{-7}
(68) Erbium	0.14	14	3	2	3	3.3	4.0	0.5	—	9×10^{-7}
(69) Thulium	0.02	1.9	0.5	0.2	0.35	0.2	0.3	0.04	—	2×10^{-7}
(70) Ytterbium	0.13	13	3	1	2.2	2.6	4.0	0.5	—	8×10^{-7}
(71) Lutetium	0.02	1.9	0.5	0.2	0.35	0.7	1.2	0.2	—	2×10^{-7}
(72) Hafnium	0.26	13	3	6	2	2.8	3.9	0.3	—	$<8 \times 10^{-6}$
(73) Tantalum	0.02	1.7	2	1.6	0.7	0.8	0.0X	0.0X	—	$<3 \times 10^{-6}$
(74) Tungsten	0.14	0.3	1	<0.8	0.45	1.8	1.6	0.6	—	$<1 \times 10^{-6}$
(75) Rhenium	0.04	—	0.001	0.0006	0.0003	—	—	—	—	8×10^{-6}
(76) Osmium	0.45	—	0.005	<0.00006	0.0003	—	—	—	—	—
(77) Iridium	0.40	—	0.001	—	0.0003	—	—	—	—	—
(78) Platinum	0.9	—	0.01	<0.01	0.02	—	—	—	—	—
(79) Gold	0.18	—	0.005	0.005	0.005	—	—	—	—	1×10^{-5}
(80) Mercury	1(?)	—	0.08	0.25	0.11	0.4	0.03	0.04	—	2×10^{-4}
(81) Thallium	0.14	—	0.5	1.3	0.13	1.4	0.8	0.0X	—	1×10^{-5}
(82) Lead	2.9	1.2	13	49	8	20	7	9	—	3×10^{-5}
(83) Bismuth	0.13	—	0.2	—	—	—	—	—	—	2×10^{-5}
(90) Thorium	0.04	2.0	7	52	2.4	12	1.7	1.7	—	4×10^{-8}
(92) Uranium	0.01	0.5	2	4	0.5	3.7	0.5	2.2	—	3×10^{-3}

*Omitting the noble gases. †Values incorporating the symbol X are known only to the order given (*i.e.*, 0.0X indicates a value in the range 0.01 to 0.09).

age of two individual rocks, a granite (G-1) and a basalt (W-1). This can be seen from Table 3, by comparing G-1 and W-1 values with those in the column headed "Earth's crust."

The figures for the specific granitic (G-1) and basaltic (W-1) rocks are included in Table 3 because they have been analyzed for practically all the elements in different geochemical laboratories throughout the world. The rock G-1 was a granite from Westerly, Rhode Island, and W-1 was a basaltic rock (specifically a diabase) from Centerville, Virginia. Several hundred kilograms of each of these rocks were crushed to a fine powder in the laboratories of the U.S. Geological Survey and samples distributed to analytical laboratories throughout the world, in order to obtain as many analyses as possible. The results were then critically examined, and the figures given in Table 3 are considered to be the best available in terms of accuracy and precision. G-1 and W-1 are undoubtedly the most thoroughly analyzed rocks and now serve as basic geochemical standards. It must be borne in mind that they are individual rocks, however, and cannot be considered to be averages of all granites and all basalts. Indeed, it is clear that G-1 is unusually rich in some trace elements (*e.g.*, thorium).

Geochemical distinction between oceanic and continental crust

Perhaps the most significant feature of the composition of the Earth's crust is that it is dominated by comparatively few elements. Only eight—oxygen, silicon, aluminum, iron, calcium, magnesium, sodium, and potassium—are present in amounts greater than 1 percent, and these eight make up almost 99 percent of the whole. Of these, oxygen comprises almost 50 percent by weight. The dominance of oxygen is even more marked when weight percentages are converted to atomic percentages, as follows: oxygen (weight percentage, 46.6; atomic percentage, 62.2), silicon (27.7; 21.2), aluminum (8.13; 6.47), iron (5.00; 1.92), calcium (3.63; 1.94), magnesium (2.09; 1.84), sodium (2.83; 2.64), potassium (2.59; 1.42). This comparison, of course, merely emphasizes the fact that the crust consists almost entirely of oxygen compounds, mostly silicates and aluminosilicates of iron, calcium, magnesium, and the alkali metals. As Goldschmidt remarked, the lithosphere may well be called the oxysphere. Clarke and his collaborators calculated that the average mineralogical composition of igneous rocks is: quartz 12.0 percent, feldspars 59.5 percent, pyroxene and hornblende 16.8 percent; biotite 3.8 percent, titanium minerals 1.5 percent, apatite 0.6 percent, and other accessory minerals 5.8 percent.

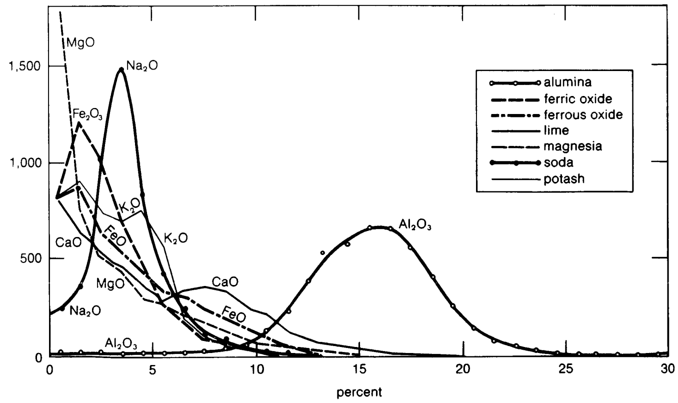


Figure 2: Frequency distribution of the major oxides in analyses of igneous rocks.

So far only the major elements have been discussed, those present in amounts greater than 1 percent by weight. Consideration must now be given to the minor and trace elements. The data are given in Table 3, and are presented diagramatically in Figure 3.

A cursory examination immediately reveals some intriguing features in the abundance pattern. The predominance of the even-numbered elements over the neighbouring odd-numbered ones is still apparent but not so regular as in the cosmic abundances. Chemical fractionations taking place during the evolution of igneous rocks from primordial matter have clearly modified this basic relationship. The odd–even relationship is most prominent in the rare-earth elements (atomic numbers 57–71), which are chemically so similar that they are little fractionated by geochemical processes.

The eight principal components of the crust

A particularly noteworthy feature is that some unfamiliar elements, such as rubidium, are relatively abundant, whereas others, such as most of the industrial metals except iron and aluminum, are actually of very low abundance. Thus boron, familiar to every homemaker in the form of borax cleansers and boric-acid antiseptics and well-known in the ancient world, is an element of extremely low abundance, much lower, for example, than zirconium, and lower even than hafnium, an element first discovered as late as 1923. Mercury, another element known to the ancient Greeks and Romans, is also of extremely low abundance. On the other hand, the so-called rare-earth elements are not extremely rare. Evidently one must make a clear distinction between abundance and availability. Some elements may be of low abundance but are readily available, because geochemical processes result in their concentration in specific deposits which can be commercially exploited. Other elements that are relatively abundant may be widely dispersed in small amounts and seldom or never occur in concentrated large deposits. A typical example is titanium, which is present in practically all igneous rocks in amounts ranging up to several percent, but which, in spite of its useful properties as a metal, is still not widely used. This is partly because of its dispersed nature and partly because of the technical difficulties in extracting the element from the minerals in which it occurs.

The distribution of minor and trace elements in igneous rocks, for those of lithophile affinity, is largely controlled by their ionic radii or size. Minor and trace elements with radii similar to those of major elements can substitute for these elements in the common minerals of the igneous rocks. The crystal structures of these minerals act as sorting mechanisms, accepting those atoms of appropriate size and rejecting others. Thus rubidium, with a radius of 1.47Å (one angstrom [Å] = 10^{-8} centimetres) is incorporated in potassium feldspar, $KAlSi_3O_8$, because its radius is close to that of potassium (1.33Å). The next higher alkali element, cesium, with a considerably larger radius (1.67Å), is not accepted into the feldspar structure; it remains in the igneous liquid during the crystallization of the major minerals until its concentration increases to such an extent that it can form the independent mineral pollucite ($CsAlSi_2O_6$).

The factor of ionic size, coupled with geochemical affinity (lithophile, chalcophile, or siderophile) is a key to the distinction between abundance and availability. Elements that are similar in size and geochemical affinity to major elements are dispersed in small amounts in common minerals; *i.e.*, rubidium in potassium feldspar, gallium in aluminum minerals, and germanium in silicate minerals. Elements that do not readily enter the common minerals of igneous rocks remain in the residual melt as crystallization proceeds. Fractional crystallization of magmas (igneous melts) normally results in a residual liquid of granitic composition. Under suitable conditions this residual liquid solidifies as a coarse-grained rock known as a pegmatite. Pegmatites are famous for their content of rare and unusual minerals, which contain many of the minor and trace elements. They are the commercial sources of lithium, beryllium, scandium, yttrium, the rare earths, cesium, niobium, and tantalum, all elements that concentrate in the residual liquid because of their specific geochemical properties.

Chalcophile elements are all of rather low abundance, and the minerals that they form, mainly sulfides and some arsenides, are not stable at the high temperatures of igneous crystallization. Sometimes these elements are found in granites and pegmatites—molybdenite (MoS_2) is a typical example. More frequently they are removed from the crystallizing magma as hot aqueous solutions and are deposited as metalliferous veins in the surrounding rock. Sometimes they may reach the surface as components in thermal springs; mercury has been deposited (as native mercury and as cinnabar, HgS) by some of these springs,

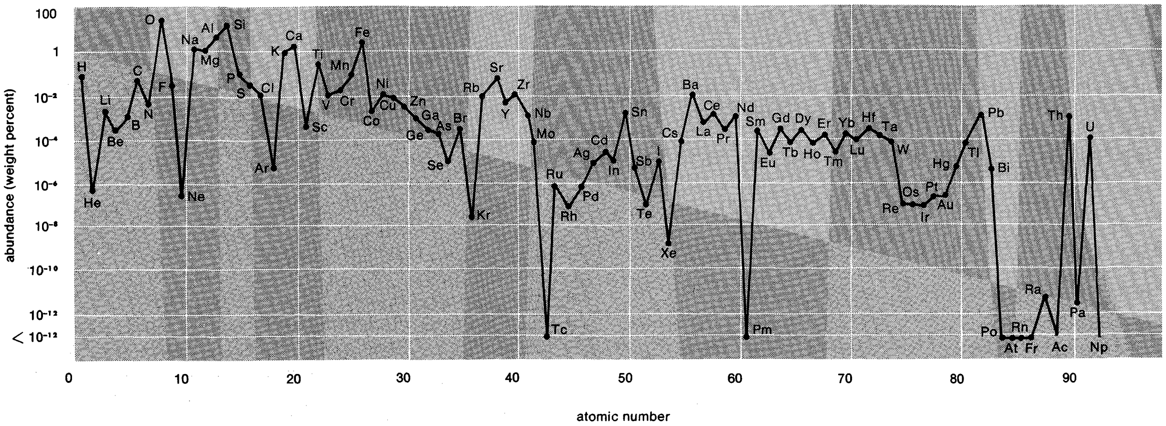


Figure 3: Crustal abundances of elements of atomic numbers 1 to 93.
From B. Mason, *Principles of Geochemistry*, 3rd ed. (1966); John Wiley & Sons, Inc.

occasionally in sufficient amounts for profitable mining.

Minor and trace elements

Sedimentary rocks. The decomposition of pre-existing rocks by weathering, the transportation and deposition of the weathering products as sediments, and the eventual formation of sedimentary rocks might be expected to produce a gross mixture of materials, thereby working against further geochemical differentiation of the elements. This is not the case; sedimentary processes frequently produce remarkable concentrations of the elements, leading to almost pure deposits of certain minerals. Some sandstones, *e.g.,* contain over 99 percent quartz and some limestones over 99 percent calcium carbonate. The ultimate is reached in salt deposits, with extensive beds of anhydrite ($CaSO_4$), gypsum ($CaSO_4 \cdot 2H_2O$), halite (NaCl), and other compounds. Goldschmidt compared the sedimentary process with a quantitative chemical analysis, involving the successive separation of specific elements or groups of elements.

Quartz (SiO_2) is highly resistant to weathering and accumulates as deposits of sand. When consolidated these deposits form sandstones, an important group of sedimentary rocks. Under special conditions almost any mineral may be deposited in sand-sized grains, but most minerals are eventually decomposed by weathering. A few resistant ones may survive and be sufficiently concentrated to form economic deposits known as placers; the most familiar are probably the gold-bearing sands, important sources of this element, but sand deposits may have economic concentrations of zirconium (as the mineral zircon, $ZrSiO_4$), titanium (as rutile, TiO_2, and ilmenite, $FeTiO_3$), tin (as cassiterite, SnO_2), and others.

The aluminosilicates of igneous rocks, mainly the feldspars, $(K,Na)AlSi_3O_8$ and $(Na,Ca)(Al,Si)_4O_8$, are relatively easily decomposed by weathering. The alkali elements and calcium are largely carried away in solution, whereas the aluminum and silicon are quickly redeposited as insoluble clay minerals. When consolidated, these minerals form shales and mudstones. The ferromagnesian minerals undergo a more complex decomposition, sometimes leading to the deposition of iron-rich sediments consisting largely of hydrated ferric oxide; such sediments are valuable iron ores in many countries.

Calcium is carried away in solution mainly as calcium bicarbonate, $Ca(HCO_3)_2$. Most of it eventually reaches the sea, where it is utilized by a vast variety of organisms as skeletal material in the form of calcite and aragonite (polymorphs—different forms—of $CaCO_3$). Accumulation of skeletal materials after death of the organisms has formed extensive deposits of limestone throughout geological time. Magnesium in seawater can react with calcium carbonate to form dolomite, $CaMg(CO_3)_2$, and in this way some magnesium is removed from solution and deposited in sediments.

Much of the magnesium, however, remains in seawater, which is essentially a dilute solution of magnesium, calcium, sodium, and potassium chlorides and sulfates, with many other elements in small amounts (Table 3). Under special geological circumstances bodies of seawater can be cut off from the open ocean, and under arid conditions the water will evaporate and extensive salt beds be deposited. Such conditions have occurred in different regions throughout geological time, and the resulting salt deposits are economically important as sources of sodium, potassium, calcium, magnesium, chlorine, and sulfur (see MINERALS AND ROCKS: *Evaporites*).

The three major groups of sedimentary rocks are sandstones, shales, and the carbonate rocks (limestones and dolomites). Much less geochemical research has been devoted to sedimentary rocks than to igneous rocks, and the data for their contents of minor and trace elements are therefore less extensive. The figures in Table 3 show that the minor and trace elements generally tend to be more concentrated in shales than in the sandstones and carbonate rocks.

Concentration of elements by sedimentary processes

The problem of arriving at an average composition for all sedimentary rocks is still largely unresolved, largely because of uncertainty in the relative amounts of shales, sandstones, and carbonate rocks. From geochemical arguments Clarke estimated the relative percentages of these three groups as 80:15:5, respectively. Actual measurements of sedimentary rocks suggest that these figures overestimate the amount of shales and underestimate that of limestones, however. Thus, a compilation of the recorded amounts of shales, sandstones, and limestones in more than 213,000 metres (700,000 feet) of sedimentary rock formations gave relative percentages of 46:32:22, respectively. The identification of a formation as a limestone, a sandstone, or a shale, however, is likely to be gross; shales usually contain considerable sand, sandstones may carry much clay, and the term limestone is applied to many rocks with 50 percent or less of carbonate. It does appear that limestones are more prominent in the geological record than might be expected from geochemical calculations, however; this probably reflects the fact that shallow-water environments are the great places of carbonate deposition, whereas the ocean deeps are the repository primarily of clay-rich sediments.

Metamorphic rocks. Comparatively few investigations have been made of the elemental composition of metamorphic rocks. Many of these rocks retain the geochemical features of their parent igneous or sedimentary materials, and their bulk composition has been little changed despite complete recrystallization and the production of new minerals and structures in some instances. Some metamorphic rocks, however, have been markedly modified by the removal of some components and the addition of others.

Average composition of sedimentary rocks

The Geological Survey of Canada has performed a comprehensive study of a large area of the Canadian Shield, a region of complex geology largely made up of metamorphic rocks. From a collection of more than 8,000 bedrock samples, the average abundances of all the major elements and a number of minor and trace elements were determined; the figures are given in Table 3. As might be anticipated, the average composition is not very different from the average composition of igneous rocks. It does show a somewhat higher silicon content, probably reflecting a preponderance of granitic over basaltic rocks and a relative abundance of quartz-rich sedimentary rocks in the original makeup of the Canadian Shield. The general validity of these abundance figures for metamorphic rocks has been confirmed by a similar study of the average composition of metamorphic rocks in the U.S.S.R., which has given closely comparable results.

Ore deposits. An ore deposit, in its simplest terms, is a portion of the Earth's crust from which some industrial raw material can be extracted at a profit. As such, its characteristics are as much economic as geochemical. Nevertheless, its formation required the operation of geochemical processes to produce the concentration of a specific element or elements in a particular place. Economics decide whether this concentration is rated as an ore deposit or merely as a deposit of scientific interest. The economics may change with time, depending upon price, availability of transportation, cost of labour, and other factors.

Table 4: Concentration Factors for Ore Bodies of the Commoner Metals

metal	percent in Earth's crust	minimum percent profitably extracted	enrichment factor necessary for an ore body
Aluminum	8.13	30	4
Iron	5.00	30	6
Manganese	0.10	35	350
Chromium	0.02	30	1,500
Copper	0.007	1	140
Nickel	0.008	1.5	175
Zinc	0.013	4	300
Tin	0.004	1	250
Lead	0.0016	4	2,500
Uranium	0.0002	0.1	500

Some general principles can, however, be enunciated. Proceeding from the average abundance of an element in the crust, and the minimum abundance that can be profitably exploited under normal circumstances, a factor of enrichment necessary to produce an ore deposit can be derived (Table 4). The economic control is immediately evident in the approximate relation between the factor of enrichment and the price of the product sought. The most extreme example of this is in diamond mining, where the product sought may be present in the rock mined in as low a concentration as 1 part in 50,000,000. Ease of extraction, of course, plays an important role in this. Diamonds are readily separated from the great mass of waste rock by a relatively simple and inexpensive process. Magnesium is commercially extracted from seawater, where its concentration is 0.13 percent, rather than from the common rock dunite, where its concentration is about 25 percent, because of the ready availability of seawater and the cheapness of the extraction process.

Ore deposits may be found in all types of rocks—igneous, sedimentary, and metamorphic—and seawater is also a significant source of such elements as sodium, chlorine, magnesium, and bromine. There are many possible processes of geochemical enrichment leading to the formation of ore deposits, and they are often the end result of a complex series of such processes acting over a long period of time. The economic importance of ore deposits has ensured their thorough study by all techniques of geological and geochemical research, but much controversy still exists regarding the origin of many of the more complex deposits.

The most readily understood ore deposits are those of sedimentary origin. They have been formed at the surface of the Earth by processes that can usually be observed operating at the present time and that can readily be simulated in the laboratory. Salt deposits are one kind whose origin is clearly amenable to such an approach. As long ago as 1849 an Italian scientist initiated laboratory studies on the evaporation of seawater and elucidated the sequence of crystallization of the different salts. Comparison of the results with the mineralogy of salt deposits revealed gross similarities but also important differences; these differences can be explained by a variety of mild metamorphic reactions resulting from burial of these deposits under overlying sediments.

The Canadian Shield

Some sedimentary deposits are not readily explicable by such an approach, however. The most extensive and economically important are the vast Precambrian iron-ore deposits, which are a major source for the hundreds of millions of tons of steel produced annually. They occur on all the continents (except perhaps Antarctica) and are uniformly of great age (about 1,900,000,000 years or older). Probably the most extensive and best exposed of these are in the Hamersley Range of Western Australia, where individual beds of iron ore are continuous over hundreds of square miles in a horizontally bedded sequence of iron ore and quartzite thousands of feet thick. The conditions that gave rise to these deposits were apparently unique to this early period in Earth history, because similar deposits are not known in younger geological formations. It has been argued that the explanation lies in an oxygen-free reducing atmosphere in early geological times, under which iron could readily be transported in solution as ferrous compounds to the ocean or large lakes, where deposition eventually took place, perhaps through the agency of primitive organisms. As soon as free oxygen appeared in the atmosphere, 1,000,000,000 to 2,000,000,000 years ago, the geochemical cycle for iron was profoundly modified, and this type of transportation and deposition ended forever.

Processes other than fractional crystallization from igneous melts also give rise to magmatic ore deposits. Economic deposits of the oxide mineral chromite ($[Fe,Mg][Cr,Al]_2O_4$), for example, occur almost entirely as bands or lenses in magnesium-rich igneous rocks. Chromite evidently crystallizes early from a magma, and, being of higher density than the liquid, it sinks to the bottom of the magma chamber and becomes concentrated as almost pure bodies of this mineral. Some accessory minerals of igneous rocks are important sources of metallic elements, but the rocks cannot be mined directly because the grade is too low. If these minerals are chemically and mechanically resistant, weathering and transportation may eventually concentrate them into workable deposits. A large proportion of the world's zirconium, hafnium, rare earths, and thorium, and some iron and titanium, come from such deposits in river and beach sands.

A large number of important ore deposits occur in metamorphic rocks. The ultimate origin of these deposits is frequently obscured by the complex processes they have undergone. If it can be established that the enclosing metamorphic rocks were of sedimentary origin, the question then arises whether the ore material was deposited along with the sediments or was introduced by circulating solutions during the metamorphism or possibly at some later time. The answer is seldom clear-cut, and such deposits continue to excite lively controversy among geochemists and economic geologists.

Mineral fuels. The mineral fuels—coal, petroleum, and natural gas—may be described as a special type of economic deposit. Geochemically they represent the concentration of carbon and hydrogen by processes that were initially biological in nature. Coal is essentially the product of accumulation of land plants in large amounts, and petroleum and natural gas are the products of marine organisms (although the origin of some petroleum and natural gas under nonmarine conditions cannot be entirely excluded). The origin of petroleum and natural gas presents a more difficult problem than coal because they are fluids and thus are free to migrate from their place of origin.

Sedimentary iron-ore deposits

The formation of coal is a relatively straightforward geochemical process that can readily be traced through its successive stages. The first requirement is a geological one—the rapid accumulation of plant material under conditions that inhibit its decomposition, followed by its burial under inorganic sediments such as shales and sandstones. The great coal-forming period in the Northern Hemisphere followed the Devonian Period (345,000,000 to 395,000,000 years ago), when abundant land plants first appeared, and has been named the Carboniferous Period (280,000,000 to 345,000,000 years ago). During this period, large areas in North America and Europe were evidently low-lying swamps that supported a lush vegetation. This vegetation died, accumulated in successive layers, and was partly decomposed by bacteria and other organisms to form peat. Burial of peat deposits under inorganic sediments brought an end to the period of bacterial decomposition, and the further changes to coal were essentially a mild metamorphism caused by an increase in temperature and pressure.

Chemically, this mild metamorphism was in large part the expulsion of carbon dioxide and water from the coal-forming substance. The main trend in the change from peat through lignite to bituminous coal and anthracite is the decrease in oxygen content and the increase in carbon (Table 5). If carried to its ultimate conclusion the product would be pure carbon in the form of graphite. This occurs comparatively rarely, but evidence for it is the presence of small amounts of graphite in many metamorphic rocks.

Table 5: Average Composition of Fuels

	carbon	hydrogen	nitrogen	oxygen
Wood	49.65	6.23	0.92	43.20
Peat	55.44	6.28	1.72	36.56
Lignite	72.95	5.24	1.31	20.50
Bituminous coal	84.24	5.55	1.52	8.69
Anthracite coal	93.50	2.81	0.97	2.72

Source: Clarke, *The Data of Geochemistry.*

Formation of coal

Coal also contains inorganic material that appears as ash when it is burned, and some coal ashes show a remarkable concentration of unusual elements. This was demonstrated by Goldschmidt in 1933, when he found appreciable amounts of germanium in some coal ashes. The Hartley Seam of the Durham Coalfield in England contains so much germanium that the ash has a brilliant yellow colour because of the presence of the oxide (GeO_2).

The source of these minor and trace elements (Table 6) and their mode of incorporation in the coal are still not fully understood. There are three possibilities: (1) these elements were taken up by the plants during growth; (2) they were carried into the coal swamp as a component of the inorganic sediments; or (3) they were absorbed during or after the coal-forming processes from circulating solutions. The first possibility is not favoured, because growing plants seldom incorporate appreciable amounts of nonorganic elements. The second possibility also is unlikely, because there is no correlation, or rather an inverse correlation, between ash content and trace element concentrations. This leaves the third possibility as the most likely one. The presence of a large amount of carbonaceous matter means that the coal-forming environment is a highly reducing one, which will favour the precipitation of some elements; the presence of hydrogen sulfide and sulfide ions will cause the precipitation of chalcophile elements (with affinity for sulfur); and complex organic compounds are noted for their absorptive, or chelating, capacity for metallic ions. Thus several individual reactions are potentially available for the fixation of foreign elements in the coal substance.

Table 6: Rare Elements in Coal Ash

	average content in coal ash (g/ton)	average content in Earth's crust (g/ton)	factor of enrichment
Boron	600	10	60
Germanium	500	1.5	330
Arsenic	500	2	250
Bismuth	20	0.2	100
Beryllium	45	2.8	16
Cobalt	300	25	12
Nickel	700	75	9
Cadmium	5	0.2	25
Lead	100	13	8
Silver	2	0.1	20
Scandium	60	22	3
Gallium	100	15	7
Molybdenum	50	1.5	30
Uranium	400	2.7	150

As pointed out above, the origin of petroleum is not as readily elucidated as the origin of coal, because petroleum can migrate from the region in which it was formed. Indeed, the very occurrence of a commercial oil field probably implies the concentration of the petroleum from a large volume of source rocks into a relatively much smaller reservoir.

The fact that petroleum is almost always found in marine sedimentary rocks has long been a basic argument in favour of a marine origin for this material. It is certainly true that some oil has been found in igneous and metamorphic rocks, but migration from a sedimentary source bed is a reasonable explanation for these occurrences. Proof of a marine origin has been forthcoming in recent years by the sensitive analyses of recent marine sediments, which show that they contain small amounts of petroleum hydrocarbons, evidently generated either directly by marine organisms or by their subsequent decomposition.

Inorganic components of coal

Natural petroleum is a complex mixture of hundreds of different hydrocarbons, but its bulk composition is remarkably constant, about 85 percent carbon and 15 percent hydrogen. It may include small amounts of organic compounds containing oxygen, sulfur, and nitrogen. Its content of other elements is exceedingly small. Petroleum ash, unlike coal ash, is not noted for its trace element content. Some petroleum ash contains appreciable amounts of vanadium, however, and has been utilized as a source of this element. A class of nitrogen-bearing organic compounds known as porphyrins include a metal atom in their molecular structure; usually this atom is iron, but other elements in this region of the periodic table, especially vanadium, nickel, and copper, may play a similar role. The vanadium content of some petroleum ash probably originates as a vanadium porphyrin in some of the organisms involved in petroleum formation.

Petroleum is always accompanied by natural gas, but many natural gas fields have no petroleum associated with them. This can probably be ascribed to greater possibility for migration for a gas as compared to a liquid. It is also possible that some natural gas is generated from coal deposits. Natural gas consists largely of methane, but small amounts of more complex hydrocarbons may be present, and it may also contain unwanted components such as nitrogen, carbon dioxide, and hydrogen sulfide. Natural gas containing hydrogen sulfide is known as "sour gas" and for long was an undesirable material because of the noxious nature of this compound; recently, however, it has been found profitable to extract the hydrogen sulfide by converting it to sulfur and then utilize the hydrocarbons.

Composition of petroleum

Natural gas is the sole source of one element, helium, the industrial demand for which has steadily increased in recent years. Comparatively few occurrences of natural gas contain sufficient helium for the extraction to be commercially profitable. Currently, the Western world's need for helium is largely met by its extraction from wells in western Texas. The explanation for this local concentration of helium in some natural gas fields is still a matter for discussion; the only reasonable source is from the disintegration of radioactive elements in the crust, but the mechanism of concentration remains something of an enigma.

Soils. Soil is a thin veneer that forms a discontinuous cover on the land areas of the Earth. Its volume and its mass are small in comparison to the major geospheres, but it is of vast importance to man. Superficially it might be considered merely as comminuted (pulverized) and decomposed bedrock; however, this viewpoint takes into account only its inorganic components and completely neglects the complex of organic compounds, living organisms, water, and included gases that gives the soil its characteristic properties and its value as the abode of life. Comminuted bedrock alone is not soil in any real sense; the Antarctic continent, where not ice-covered, has a surface of comminuted bedrock, as does the Moon, but neither material can properly be termed soil.

Soils result from the weathering of rocks, and hence their composition might be expected to reflect the composition of the rocks from which they were formed. This is true only in a very broad sense, however. Environmental factors play an important part in soil formation. The same parent rock may give rise to very different soils under different conditions. Climate, topography, vegetation, biological activity, and time are all important factors in determining the nature of a soil. Climate is probably the most important of these, as can be demonstrated by contrasting the soils developed on the same rock type under tropical and temperate conditions. In general, the soil in the humid tropics will be different in texture and composition and much less fertile, as a result of the intense leaching brought about by high rainfall, high temperatures, and the almost complete removal of organic matter by microorganisms.

The complex of inorganic compounds, organic compounds, water, and air that makes up the soil is in a continual state of change. Water tends to dissolve and remove the relatively soluble elements such as calcium, magnesium, sodium, and potassium, and the comparatively insoluble elements—aluminum, iron, and silicon—are thereby relatively enriched in the soil. The enrichment of iron is frequently manifested by a red-brown or yellow-brown colour caused by an accumulation of iron oxides. The most reactive part of the soil is the complex of clay minerals and organic matter, which is largely responsible for its agronomic characteristics. True soil does not exist without the presence of colloidal and organic matter. The relative absence of soils in desert areas reflects the fact that chemical and physical weathering of rocks alone does not necessarily result in soil formation. Most soil processes are directly or indirectly biological in nature. Organisms and organic compounds produced by their vital activities or their decomposition are effective agents for dissolving and extracting many elements from the inorganic constituents of the soil, thereby making them available for plant growth.

Helium in natural gas

Although soils do differ in composition, the range of variation in the major elements is rather small. Minor and trace elements may show considerably greater variability, as is illustrated by Figure 4. The importance of certain trace elements in the soil for the healthy growth of plants, and through the plants, of the animals that graze on them, has become increasingly apparent in recent years. Most soils contain these trace elements in sufficient amounts, but when deficiencies are present, puzzling diseases appear, which in the past have rendered large areas of otherwise suitable land unavailable for farming. On a large area in the North Island of New Zealand, for example, although it grew satisfactory pasture, sheep and cattle failed to thrive and eventually died if not removed. As a result, much of this area was given over to afforestation. It was eventually discovered that cobalt, in the amount of a few parts per million, would completely eliminate the disease when applied in fertilizer or administered directly to the animals. The ultimate explanation is the need of animals (but not plants) for vitamin B_{12}, which contains an atom of cobalt in its structure.

From *Chemistry of the Soil* by Firman Bear. Copyright 1964 by Litton Educational Publishing,Inc.

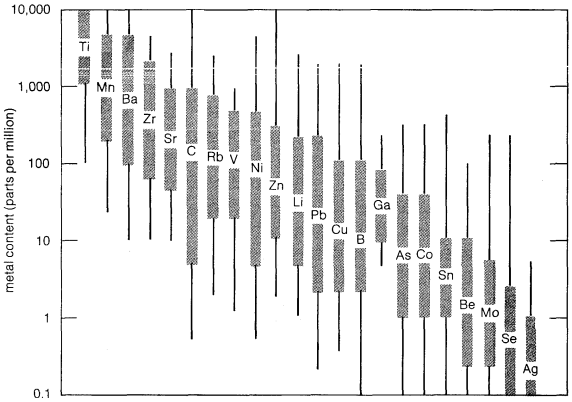


Figure 4: Range of the content of some minor elements in normal soils. Thin lines indicate ranges of more unusual values.

Occasionally, an excess of a specific element may have a deleterious effect on plant growth. Most obvious, of course, are the alkaline or saline soils of desert and coastal areas on which only an impoverished vegetation exists. Magnesium-rich soils are notably infertile; such soils develop on areas of ultrabasic igneous rocks consisting largely of olivine, $(Mg,Fe)_2SiO_4$, and the boundaries of these areas can frequently be readily mapped from aerial photographs by the marked change in vegetation. Sometimes plants take up available trace elements in amounts deleterious to animals grazing on them. A well-known example is *Astragalus racemosus* (locoweed), which in some areas of the western U.S. contains sufficient selenium to be poisonous to grazing animals.

The possible correlation between soil geochemistry and the geographical distribution of disease is thus a field of extreme significance which as yet has been insufficiently studied. The problem is a complex one, in large part because of the difficulty in isolating the numerous factors involved.

THE HYDROSPHERE

The hydrosphere is the discontinuous shell of water—fresh, salt, and solid—on the surface of the Earth. As such it comprises the oceans and the connecting seas and inlets, the lakes, rivers, and streams, the groundwater that feeds them, and the snow and ice cover of high altitudes and high latitudes. The mass of the ocean waters far outweighs the other parts of the hydrosphere. Goldschmidt estimated that there are 273 litres of water in all its forms for every square centimetre of the Earth's surface made up as follows:

	Litres	Kilograms	Total mass
Seawater	268.4	278.1	1.4×10^{18} tons
Freshwater	0.1	0.1	5.1×10^{14} tons
Continental ice	4.5	4.5	2.3×10^{16} tons

Seawater thus makes up over 98 percent of the total mass of the hydrosphere, and its composition (Table 3) essentially can be taken as giving the average composition of the hydrosphere.

Research during the past century has demonstrated that the composition of seawater is essentially uniform and that the relative proportions of the various ions are practically constant. In the open ocean the salinity (approximately the total weight of dissolved solids per kilogram) averages about 35 parts per thousand, but may rise to 40 parts per thousand in regions such as the Red Sea and the Persian Gulf, where rainfall and inflow are low and evaporation high. Sodium chloride is the dominant compound of the salts in solution and comprises about three-quarters of the whole; the remainder consists largely of chlorides and sulfates of magnesium, calcium, and potassium.

Minor and trace elements in soils

Though many data on minor and trace elements in seawater (Table 3) are available, the interpretation of these data is subject to some uncertainties. The concentrations of these elements are probably more variable than for the major elements and may depend to some degree on the sampling location. This is particularly true for the elements that are utilized by marine organisms. Phosphorus is a good example; it is markedly depleted in surface and near-surface waters by biological activity, but it enriches the deeper parts of the ocean through the dissolution of dead organisms. Silica is brought into the ocean in large amounts in solution in river water, but most of it is soon removed to become the skeletal material of diatoms, radiolaria, and sponges.

Seawater also contains dissolved gases in variable amounts. Seawater of normal salinity at 0° C (32° F) in equilibrium with the atmosphere will contain about eight millilitres per litre of dissolved oxygen and 14 millilitres per litre of dissolved nitrogen. Dissolved nitrogen is essentially an inert constituent, but dissolved oxygen plays a fundamental role in the growth and decay of organisms and so varies greatly in concentration from place to place. In stagnant regions that are rich in decaying organic matter the water may be completely depleted in free oxygen, and a considerable concentration of hydrogen sulfide may be present. Much of the Black Sea below a depth of a few hundred metres is in this condition.

Another dissolved gas of prime importance for biological activity is carbon dioxide. The conditions controlling its concentration are quite complex, however, because in solution it can be present as free carbon dioxide, as undissociated carbonic acid, as carbonate ions, and as bicarbonate ions. The concentration of these ions will also be affected by biological activity and by the precipitation or dissolution of calcium carbonate.

The circulation of water through the hydrosphere is controlled in large part by the reservoir effect of the oceans. Evaporation from the ocean surface is precipitated as rainfall. Of that falling on the land, some is directly reevaporated, some is absorbed into the reservoir of groundwater, and some flows off directly into rivers and streams. The total annual rainfall on the Earth is estimated to be 123×10^{18} grams, and the total annual runoff to the oceans 32×10^{18} grams.

Even the purest rainfall contains some material in solution, not only dissolved gases but also nonvolatile material. Rainfall near seacoasts always contains some sodium chloride and small amounts of other marine salts, the concentration of which falls off generally with distance from the ocean. Rainfall in industrial regions may of course contain a variety of pollutants; in many areas it is essentially a dilute sulfuric acid solution. Such material may also be carried far beyond the place of origin; acid rainfall in the Scandinavian countries probably originates in part from England and Germany.

Chemical composition of seawater

The runoff from the land contains additional material in solution, picked up during its circulation through the crustal rocks. River water averages about 120 parts per million dissolved solids, but the range is great, from about 10 parts per million up to several thousand parts per million. Commonly, the range is from 50 to 200 parts per million; contents greater than 200 parts per million are usually the result of human activities or of drainage from soils containing soluble salts, as in desert regions.

With an average content of 120 parts per million dissolved matter, the rivers of the world deliver 3.9×10^9 tons of material in solution to the sea each year. The average concentration of the important constituents (in parts per million) is: bicarbonate, 58.4; sulfate, 11.2; chloride, 7.8; nitrate, 1.0; calcium, 15.0; magnesium, 4.1; sodium, 6.3; potassium, 2.3; iron, 0.67; and silica, 13.1. Although these ten constituents account for most of the dissolved material, many other elements have been detected in river and lake waters. Comparatively little information is available for these elements, but the results of an extensive study of the large rivers of North America is summarized in Figure 5. The figures are in micrograms per litre—that is, in parts per thousand million.

From W.H. Durum and Joseph Haffty, *Geochimica et Cosmochimica, Acta*, vol. 27 (1963); Pergamon Press

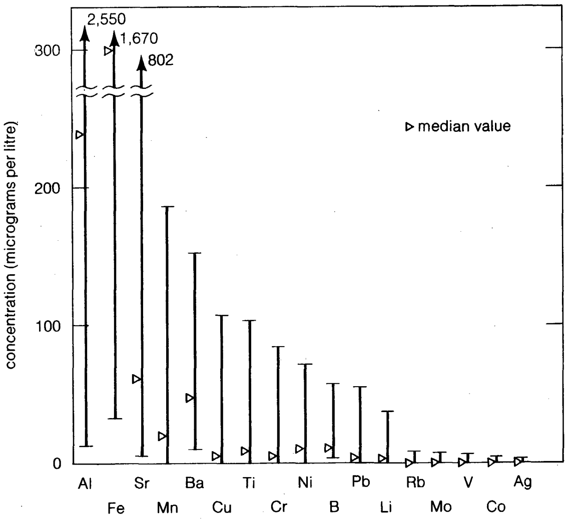


Figure 5: Concentrations of minor and trace elements in river waters of North America.

The 3.9×10^9 tons carried annually in solution to the oceans are but a small fraction of the total amount of material in solution in the oceans. Nevertheless, when integrated over the whole of geological time, more than 4×10^9 years, it greatly exceeds the present material in solution. Some of the material, especially sodium chloride, is of course cyclical, being circulated from the oceans to the land as aerosols and incorporated in marine sedimentary rocks and ultimately in large part being returned to the oceans in runoff.

Table 7: Geochemical Balance of Some Elements in Seawater

element	potential amount supplied to oceans (g/ton)	amount present in seawater (g/ton)	percentage in solution
Lithium	39	0.17	0.4
Boron	2	4.5	250
Fluorine	540	1.3	0.2
Sodium	16,980	10,800	64
Magnesium	12,540	1,290	10
Phosphorus	708	0.09	0.01
Sulfur	312	904	290
Chlorine	188	19,400	10,300
Potassium	15,540	392	2.5
Calcium	21,780	411	1.9
Arsenic	3	0.003	0.1
Bromine	0.97	67	6,900
Rubidium	186	0.12	0.06
Strontium	180	8.1	4.6
Iodine	0.18	0.06	33
Cesium	4	0.0003	0.008
Barium	150	0.02	0.01

Goldschmidt made an interesting calculation on the geochemical balance in seawater. From the amount and composition of sedimentary rocks he estimated that erosion during geological time had amounted to about 160 kilograms of igneous rock per square centimetre of the Earth's surface. Combining this figure with the amount of seawater per square centimetre, 273 kilograms, he derived a figure of 600 grams of igneous rock eroded per kilogram of seawater. Assuming this 600 grams had gone fully into solution (obviously a gross simplification but a limiting one), he drew up a balance sheet between the amounts of different elements potentially supplied to the oceans and the amounts actually present. Some of these figures are presented in Table 7. Despite the imperfections of the method, the results are certainly significant in a qualitative sense. Some elements—chlorine, bromine, boron, and sulfur—are present in seawater in amounts far in excess of those that can have been derived by erosion. The source of these "superabundant" elements has probably been volcanism and related magmatic activity. Halides, sulfates, and borates are deposited by volcanic gases and carried in solution in hot springs. The relative depletion of fluorine with respect to chlorine in seawater can be ascribed to the precipitation of highly insoluble fluorine-bearing compounds, mainly apatite (calcium fluophosphate). Sodium clearly remains in solution to a much greater extent than potassium; the latter element reacts with sedimentary materials to form insoluble potassium-bearing silicates such as illite and glauconite, which have no sodium-bearing analogs. Calcium is removed from solution much more effectively than strontium, evidently because it is utilized by organisms. Goldschmidt pointed out that many highly poisonous elements, such as arsenic and selenium, have been potentially supplied in dangerous amounts. Their concentration remains very low, however, presumably because of efficient processes of removal as insoluble compounds. Adsorption on colloidal particles of clay and iron oxides is a likely process.

Steady state of oceanic composition

The geological and geochemical evidence indicates that the ocean waters are, and have been for a long time, in a steady state of essentially unchanging composition. The addition of material by runoff from the land is adjusted by reactions within the ocean waters or between the ocean waters and sedimentary materials whereby the concentrations of the individual elements remain essentially constant. How far back in geological time this steady state has persisted remains something of an open question. The existence of most forms of marine life from the Cambrian to the present indicates a uniformity of marine conditions over the past 600,000,000 years; how far back into the Precambrian this uniformity extended is more difficult to elucidate. The earlier discussion of Precambrian iron formations suggested the possibility of very different atmospheric composition some 2,000,000,000 years ago, and the considerable interdependence of atmospheric and oceanic composition indicates that this might have resulted in marked geochemical differences in the ocean waters.

THE ATMOSPHERE

The atmosphere is the most homogeneous and thus the most easily studied of the geospheres. Its mass is readily determined from the product of the average height of the mercury barometer in centimetres, the density of mercury (13.6 grams per cubic centimetre), and the area of the Earth (5.1×10^{18} square centimetres). Recent calculations give 51.17×10^{20} grams for its total mass.

The composition is also relatively simple, although a considerable number of gases may be present in small amounts (Table 8). Almost 99 percent consists of oxygen and nitrogen, with argon making up most of the remainder. Carbon dioxide, essential for plant life, is present in an extremely small amount. Some gases not listed in Table 8 may be present as local or even regional pollutants—city dwellers are becoming increasingly aware of oxides of sulfur as atmospheric pollutants, and the scientific study of smog is largely concerned with reactions taking place between hydrocarbons, oxides of nitrogen, oxygen and ozone.

Table 8: Average Composition of the Atmosphere

gas	composition by volume (ppm)*	composition by weight (ppm)*	total mass ($\times 10^{20}$ g)
Nitrogen	780,900	755,100	38.648
Oxygen	209,500	231,500	11.841
Argon	9,300	12,800	0.655
Carbon dioxide	300	460	0.0233
Neon	18	12.5	0.000636
Helium	5.2	0.72	0.000037
Methane	1.5	0.94	0.000043
Krypton	1	2.9	0.000146
Nitrous oxide	0.5	0.8	0.000040
Hydrogen	0.5	0.035	0.000002
Ozone†	0.4	0.7	0.000035
Xenon	0.08	0.36	0.000018

*ppm = parts per million. †Variable, increases with height.

The atmosphere gradually thins out into the vacuum of outer space, and its upper limit can conveniently be placed at about 600 kilometres. An important zone in the stratosphere is known as the ozonosphere, a diffuse layer characterized by an increase in the concentration of ozone, O_3. This zone is highly important for life on Earth because it absorbs most of the ultraviolet radiation from the Sun; if this penetrated to the Earth's surface it would act as a potent sterilizer, fatal for most forms of life. It also helps to maintain a more uniform surface temperature by reducing the loss of heat by radiation to space—the so-called greenhouse effect.

The geochemical history of the atmosphere has been a complex one. Scientists agree that the present atmosphere is quite different from the original one. It is certainly quite different from those of the other planets. It is reasonable to conclude that this reflects, in part at least, the Earth as the abode of life. The Earth's atmosphere differs from those of its neighbours in the solar system probably in large part through the action of photosynthesis, a complex biological process which was probably preceded by a lengthy period of organic evolution.

The nature of the Earth's primitive atmosphere is still a subject of some speculation. Some scientists, reasoning by analogy with the larger planets such as Jupiter, have argued for an original atmosphere consisting largely of methane and ammonia. Others have considered that present-day volcanic gases may indicate the nature of the primitive atmosphere, in which case it contained carbon dioxide, possibly carbon monoxide, nitrogen, and water vapour. In either case, free oxygen was absent. If the evolution of the atmosphere is traced backward in time through the geological record, then extensive terrestrial photosynthesis is indicated by an abundance of land plants in Devonian times, about 400,000,000 years ago. Marine photosynthesis, however, is much older, since practically all the major groups of marine organisms were established by the beginning of the Cambrian period, some 570,000,000 years ago. As discussed earlier, the extensive Precambrian iron formations suggest an oxygen-free atmosphere which was terminated about 2,000,000,000 years ago. This evidence has been translated into estimates of oxygen content of the atmosphere of about 1 percent of the present level 2,000,000,000 years ago, about 10 percent of the present level at the beginning of the Cambrian Period, and essentially the present content by Devonian times.

Although it is not yet possible to know the quantitative composition of the primitive atmosphere, the geochemical processes that have operated to modify its composition during geological time can be evaluated. These processes can be summarized as a series of gains and losses. Additions to the atmosphere comprise: (1) gases released by igneous activity; (2) oxygen and hydrogen produced by the photochemical dissociation of water vapour; (3) oxygen produced by photosynthesis; (4) helium produced by the radioactive breakdown of uranium and thorium; and (5) argon produced by the radioactive breakdown of potassium. Atmospheric losses include: (1) oxygen removal by oxidation of ferrous to ferric iron, sulfur compounds to sulfates, hydrogen to water, and similar reactions; (2) carbon dioxide removed by the formation of coal, petroleum, and the death and burial of organisms; (3) carbon dioxide removed by the formation of calcium and magnesium carbonates; (4) nitrogen removed by the formation of oxides of nitrogen in the air and by the action of nitrifying bacteria in the soil; and (5) hydrogen and helium by escape from the Earth's gravitational field.

Photosynthesis has certainly been the most significant process in controlling atmospheric composition during much of geological time. Through this process carbon dioxide and water are converted to carbohydrate, with the accompanying release of oxygen. Much of this carbohydrate is consumed by animals and reconverted to carbon dioxide and water by respiration, and oxidative decay leads to the same result. Some, however, is incorporated into sediments; part may go to form exploitable deposits of coal and petroleum, but most of it remains as disseminated carbonaceous material; the average carbon content of sedimentary rocks is about 0.4 percent.

Removal of carbon dioxide

Quantitatively, more significant amounts of carbon dioxide have been removed from the atmosphere in the form of limestone and dolomite. Most of this removal has been effected by marine organisms, especially algae and corals, but direct inorganic precipitation may occur, especially in warm tropical waters. Judging from the vast deposits of limestone and dolomite throughout the sedimentary record, this process has operated with a remarkable degree of uniformity throughout geological time. Extensive limestone formations are possibly less common in older Precambrian rocks, indicating a slower beginning of carbonate precipitation. It is truly remarkable that so much carbonate rock has been deposited during geological time, with the carbonate being ultimately derived from an atmosphere which may never have contained a much higher concentration of carbon dioxide than is present today. It has been pointed out that reactions like the decomposition of calcium silicate—$CaSiO_3 + CO_2 = CaCO_3 + SiO_2$, in which $CaSiO_3$ is calcium silicate, CO_2 is carbon dioxide, $CaCO_3$ is calcium carbonate, and SiO_2 is silicon dioxide, which tends to go toward the right at ordinary temperatures—will act as buffering mechanisms to keep the carbon dioxide concentration of the atmosphere at a continuously low level.

If the carbon dioxide concentration has remained essentially constant, and yet this compound has been continuously extracted to form carbonates and organic compounds, then clearly a balancing source of "new" carbon dioxide is required. This evidently has been provided by volcanism and other igneous activity. The Earth is steadily being degassed, in the sense that gaseous compounds contained in the mantle are escaping to the surface. The presence of carbon dioxide in the mantle has been demonstrated by the presence of microscopic inclusions of liquid carbon dioxide in the minerals of the peridotite xenoliths (rocks contained within other rocks) brought up in some volcanoes. Along with carbon dioxide, much water and small amounts of other volatiles are being added continuously from sources in the mantle. Ultimately, the hydrosphere, as well as the atmosphere, is the product of the degassing of the Earth's interior.

Of the remaining atmospheric gases, argon presents some intriguing features. Argon is by far the most abundant of the inert gases on Earth, whereas in the universe as a whole it is much less abundant than either helium or neon. In addition, its isotopic composition is quite distinct, consisting almost entirely of argon-40, whereas in the Cosmos argon-36 is the most abundant isotope. The reason for these anomalies is that atmospheric argon is almost entirely radiogenic, the product of the decay of the potassium-40 isotope of potassium.

Similarly, the helium in the atmosphere is probably entirely the product of the radioactive decay of uranium and thorium. Actually, the atmosphere contains only about 10 percent of the total amount of helium generated from these sources during geological time. Some of this helium remains occluded in the rocks where it was formed, some has escaped from the upper atmosphere. Helium (and hydrogen), consisting of light atoms, can escape from the gravitational field of the Earth, whereas heavier gases cannot. A minor source of atmospheric oxygen throughout geological time is probably the photochemical decomposition of water vapour in the upper atmosphere, with the subsequent loss of the hydrogen to outer space.

Some oxygen has been removed from the atmosphere by oxidative reactions, of which the most significant has been the conversion of ferrous to ferric iron. In igneous rocks the average ferrous-to-ferric iron ratio (FeO/Fe_2O_3) is greater than unity, whereas in sedimentary rocks the proportion is reversed, ferric iron being dominant over ferrous iron. Other oxidative reactions are the conversion of manganous compounds to manganese dioxide and of hydrogen sulfide to free sulfur and sulfate. Nitrogen is almost inert geochemically, but a little is fixed as oxides of nitrogen by lightning, and somewhat more by the action of nitrifying bacteria in the soil. Most of this nitrogen is ultimately returned to the atmosphere by the decay of the organisms. Oxides of nitrogen formed in the atmosphere are removed in rain as nitrite and nitrate. Nitrogen does not accumulate in the soil, however, except perhaps under extremely arid conditions, as in the deserts of northern Chile, the locale of the unique nitrate deposits.

The geochemical cycle

The preceding discussion has shown that at the time of formation of the Earth the chemical elements already had been considerably fractionated: the universe consists almost entirely of hydrogen and helium, probably with less than 1 percent of the heavier elements. The Earth, on the other hand, consists almost entirely of the heavier elements. Helium is a very rare element on Earth, so rare in fact that it was first discovered as an unidentified line in the Sun's spectrum in 1868, some 30 years before it was detected on Earth. Hydrogen is moderately abundant on Earth, largely because it combines with oxygen to form water, whereas helium is an inert element.

Significance of inert gases

The rarity of helium and the other inert gases neon, krypton, and xenon on Earth is good evidence that the Earth formed by the accretion of small solid objects, or planetesimals. (Argon is a special case, since most of the Earth's argon has been formed within the planet by the radioactive decay of potassium.) These planetesimals had no atmosphere, and the atmosphere of the Earth has been derived by the outgassing of combined and occluded gases within these planetesimals. This process has operated throughout geological history and is probably still continuing; volcanic activity not only brings up solid material from the Earth's interior but also large amounts of gases, principally water vapour, carbon monoxide and dioxide, and nitrogen. The oxygen in the present atmosphere is almost entirely the product of photosynthesis, whereby carbon dioxide and water are converted to carbohydrate and free oxygen.

Direct information on the composition of the Earth's crust is available in the form of thousands of analyses of individual rocks, the average of which provides a reasonably precise estimate of the bulk composition. For the mantle and the core the information is indirect and thus much less precise. The origin of the Earth by the accretion of planetesimals is a well-founded hypothesis, however, and meteorites are probably examples of planetesimals that have survived from the preplanetary stage of the solar system. It thus seems likely that the Earth formed by the accretion of solid bodies with the average composition of stony meteorites. The accretion process, however, led to massive segregation of the elements. Much of the iron was reduced to the metallic state and sank to the centre to form the core, carrying with it the major part of the siderophile elements. Lithophile elements, those with a greater affinity for oxygen than iron, combined as oxide compounds, mostly silicates, and provided material for the mantle and crust. Chalcophile elements would tend to form sulfides; however, few sulfides are stable at the high temperatures of the Earth's interior, so the fate of the chalcophile elements during the early history of the Earth is somewhat uncertain.

This primary geochemical differentiation of the Earth can be interpreted in terms of the system iron–magnesium–silicon–oxygen–sulfur, because these five elements make up about 95 percent of the Earth. There was insufficient oxygen to combine with the major metallic elements iron, magnesium, and silicon; because magnesium and silicon have a greater affinity for oxygen than iron, these elements combined completely with oxygen, and the remaining oxygen combined with part of the iron, leaving the remainder as the metal iron and iron sulfide. As indicated above, the metal sank to form the core, carrying with it the major part of the siderophile elements. The iron sulfide phase probably incorporated much of the chalcophile elements; it does not seem to have formed a distinct shell within the Earth and probably remains primarily in disseminated form through the mantle and the core.

This primary geochemical differentiation was largely controlled by two independent factors, the geochemical affinity of the individual elements and the density of the principal phases. It is significant that the density of the individual elements is irrelevant in this context. Uranium and thorium, for example, are very heavy elements; nevertheless, they are concentrated in the crust, not in the core, because of their lithophile character (affinity for oxygen).

Separation of crust from mantle

The separation of crust from mantle has involved rather different geochemical processes than those of the segregation of the core. The core represents the physical separation of a high-density metallic liquid immiscible with the silicate material of the mantle and crust, whereas the separation of crust from mantle has resulted from more subtle geochemical processes. The mantle is made up very largely of the magnesium-iron silicates olivine and pyroxene (or their high-pressure equivalents). As discussed earlier, these minerals act as sorting mechanisms for other elements, accepting those of ionic size comparable to magnesium and iron and rejecting those much larger or smaller. The crust differs from the mantle in being enriched in those elements that do not readily enter the crystal structures of olivine and pyroxene. The common elements include the alkali metals (except lithium), aluminum, and calcium to some degree. These are combined in the crust in the form of the feldspar minerals, and one of the principal differences between crust and mantle is the dominance of feldspar in the former and its absence in the latter.

The differentiation of the crust from the mantle and core permits consideration of the outer part of the Earth as a distinct geochemical system. The distribution and migration of the elements within the crust (and in the hydrosphere and atmosphere) are the results of processes of magmatism, sedimentation, and metamorphism. The fate of an element during magmatic crystallization is primarily a function of its ionic size. If of appropriate size it enters the structure of one or other of the major minerals; otherwise, it tends to remain in the magmatic liquid until its concentration increases sufficiently for it to form specific minerals in which it is a major component. Volatile elements and compounds in the magma (principally water, carbon dioxide, nitrogen, and sulfur compounds) escape and are added to the atmosphere and hydrosphere.

Sedimentary processes lead to further geochemical differentiation. Liquid water is the medium in which these processes largely operate, and the solution chemistry of the individual elements is the key to their geochemical behaviour. The atmosphere and hydrosphere interact importantly with the lithosphere. Water and carbon dioxide are incorporated in sedimentary minerals; soluble ions, especially sodium, are added to the hydrosphere. Processes involving living organisms are intimately associated with sedimentation; indeed, some abundant and widespread sedimentary rocks, such as coal and limestone, are essentially biological in origin. These biological processes thus have had significant geochemical results. Photosynthesis also has been responsible for the present composition of the atmosphere.

Effects of metamorphism

Metamorphism tends in some degree to reverse the geochemical differentiation of the elements produced by igneous and sedimentary processes. In general, metamorphism promotes a more uniform distribution of the elements and tends to erase compositional differences within large rock masses. Unlimited metamorphism would result in a situation in which the whole lithosphere reached a uniform composition; such a trend is exemplified by the comparatively monotonous chemical and mineralogical composition of many extensive areas of metamorphic rocks.

This overall picture of the migration of the elements in the outer part of the Earth is the basis for the concept of the geochemical cycle (Figure 6). The geochemical cycle may be considered as beginning with the crystallization of a magma, which may produce a variety of igneous rocks and which adds volatile materials to the atmosphere and hydrosphere. Weathering decomposes the igneous rocks, and erosion eventually transports the decomposed rock to the oceans, partly as solid matter to form sedimentary rocks and partly as material in solution. Crustal movements may transport sedimentary and igneous rocks to deeper levels in the crust, where the combined effects of heat and pressure induce metamorphism. If the temperature rises above the melting point of some of the rock material, magma is regenerated, and the cycle is complete. Such a geochemical cycle seldom goes to completion; at some stage it may be indefinitely halted or its direction reversed. Nor is the cycle a closed one, either materially or energetically. The crust receives "primary" magma generated in the upper mantle, bringing with it additional material and energy in the form of heat. The surface of the Earth receives an increment of meteoritic matter from outer space, insignificant in amount but nevertheless detectable in deep-sea deposits, where the rate of sedimentation is very low.

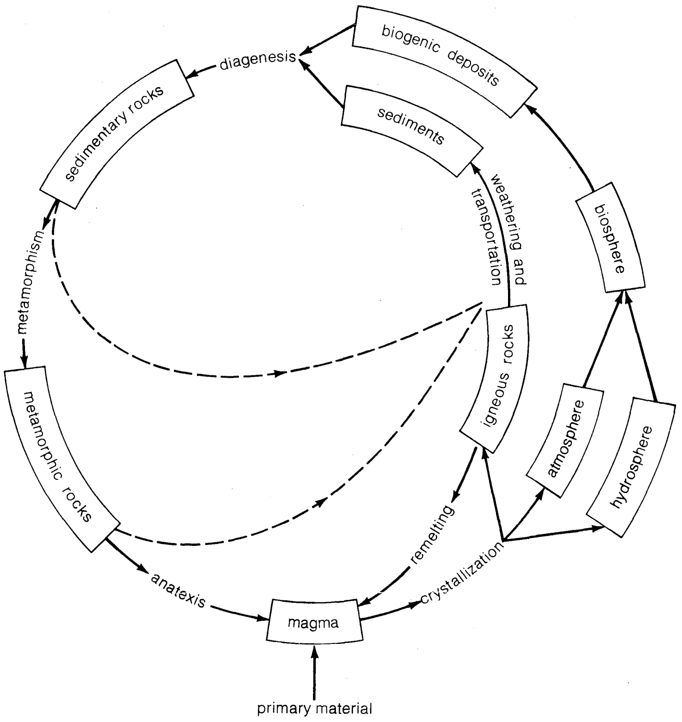


Figure 6: The geochemical cycle.

From B. Mason, *Principles of Geochemistry*, 3rd ed. (1966); John Wiley & Sons, Inc.

The geochemical cycle provides a useful conceptual framework for considering the course followed by an individual element in proceeding through the different stages. For a specific element, a complete understanding of its behaviour throughout the cycle is one of the major objectives of geochemical research. An element may tend to concentrate in a specific type of deposit at a particular stage in the cycle, or it may remain dispersed throughout the entire cycle. The geochemical cycles of a few elements have been elucidated in considerable detail, but for many of them knowledge is still fragmentary. (B.H.M./Ed.)

PERIODIC LAW AND TABLE

Dmitry I. Mendeleyev discovered in the mid-19th century that the chemical elements show a periodic recurrence of properties when they are arranged in a certain order, which is approximately the order of increasing atomic weight. The statement of this fact, which has been of inestimable value in the development of chemistry, is called the periodic law; a tabular arrangement of the elements that brings those with similar properties into juxtaposition is called the periodic table or periodic system of the elements.

It was recognized during the second decade of the 20th century that the order of elements in the periodic system is that of their atomic numbers, the integers which are equal to the positive electrical charges of the atomic nuclei expressed in electronic units; and in subsequent years great progress was made in explaining the periodic law in terms of the electronic structure of atoms and molecules. This clarification has increased the value of the law, which is used as much today as it was at the beginning of the 20th century, when it expressed the only known relationship among the elements.

History of the periodic law

The early years of the 19th century witnessed a rapid development in analytical chemistry—the art of distinguishing different chemical substances—and the consequent building up of a vast body of knowledge of the chemical and physical properties of both elements and compounds. This rapid expansion of chemical knowledge soon necessitated classification, for on the classification of chemical knowledge are based not only the systematized literature of chemistry but also the laboratory arts by which chemistry is passed on as a living science from one generation of chemists to another. Relationships were discerned more readily among the compounds than among the elements; it thus occurred that the classification of elements lagged many years behind that of compounds. In fact, no general agreement had been reached among chemists as to the classification of elements for nearly half a century after the systems of classification of compounds had become established in general use.

Triads of elements

J.W. Döbereiner in 1817 showed that the combining weight, meaning atomic weight, of strontium lies midway between those of calcium and barium; and some years later he showed that other such "triads" exist (chlorine, bromine, and iodine [halogens] and lithium, sodium, and potassium [alkali metals]). J.-B.-A. Dumas, L. Gmelin, E. Lenssen, Max von Pettenkofer, and J.P. Cooke expanded Döbereiner's suggestions between 1827 and 1858 by showing that similar relationships extended further than the triads of elements, fluorine being added to the halogens, and magnesium to the alkaline-earth metals; while oxygen, sulfur, selenium, and tellurium were classed as one family, and nitrogen, phosphorus, arsenic, antimony, and bismuth as another family of elements.

Attempts were made subsequently to show that the atomic weights of the elements could be expressed by an arithmetic function, and in 1862 A.-E.-B. de Chancourtois proposed a classification of the elements based on the new values of atomic weights given by Stanislao Cannizzaro's system of 1858. De Chancourtois plotted the atomic weights on the surface of a cylinder with a circumference of 16 units, corresponding to the approximate atomic weight of oxygen. The resulting helical curve brought closely related elements onto corresponding points above or below one another on the cylinder, and he suggested in consequence that "the properties of the elements are the properties of numbers," a remarkable prediction in the light of modern knowledge.

CLASSIFICATION OF THE ELEMENTS

Several years later, in 1864, J.A.R. Newlands proposed a system of classification of the elements in the order of increasing atomic weights, the elements being assigned ordinal numbers from unity upward and divided into seven groups having properties closely related to the first seven of the elements then known: hydrogen, lithium, beryllium, boron, carbon, nitrogen, and oxygen. This relationship was termed the law of octaves, by analogy with the seven intervals of the musical scale.

Then in 1869, as a result of an extensive correlation of the properties and the atomic weights of the elements, with special attention to valency (that is, the number of single bonds the element can form), Mendeleyev proposed the periodic law, by which "the elements arranged according to the magnitude of atomic weights show a periodic change of properties." Lothar Meyer had independently reached a similar conclusion, published after the appearance of Mendeleyev's paper.

The first periodic table. Mendeleyev's periodic table of 1869 contained 17 columns, with two nearly complete periods (sequences) of elements, from potassium to bromine and rubidium to iodine (see Figures 8 and 9), preceded by two partial periods of seven elements each (lithium to fluorine and sodium to chlorine), and followed by three incomplete periods. In an 1871 paper Mendeleyev presented a revision of the 17-group table, the principal improvement being the correct repositioning of 17 elements. He, as well as Lothar Meyer, also proposed a table with eight columns obtained by splitting each of the long periods into a period of seven, an eighth group containing the three central elements (such as iron, cobalt, nickel; Mendeleyev also included copper, instead of placing it in Group I), and a second period of seven. The first and second periods of seven were later distinguished by use of the letters "a" and "b" attached to the group symbols, which were the Roman numerals.

Addition of a new "zero" group

With the discovery of the noble gases helium, neon, argon, krypton, radon, and xenon by Lord Rayleigh (John William Strutt) and Sir William Ramsay in 1894 and the following years, Mendeleyev and others proposed that a new "zero" group to accommodate them be added to the periodic table. The "short-period" form of the periodic table, Figure 7, with Groups 0, I, II, . . . VIII, became popular, and remained in general use until about 1930.

Based upon an earlier (1882) model of T. Bayley, J. Thomsen in 1895 devised a new table. This was interpreted in terms of the electronic structure of atoms by Niels Bohr in 1922. In this table, Figure 8, there are periods of increasing length between the noble gases; the table thus contains a period of 2 elements, two of 8 elements, two of 18 elements, one of 32 elements, and an incomplete period. The elements in each period may be connected by tie lines with one or more elements in the following period. The principal disadvantage of this table is the large space required by the period of 32 elements and the difficulty of tracing a sequence of closely similar elements. A very satisfactory compromise, shown in Figure 9, is to compress the period of 32 elements into 18 spaces by omitting the 14 lanthanons (lanthanides) and the 14 actinons (actinides) and listing them in a special double row below the other periods.

Other versions of the periodic table. Alternate long forms of the periodic table have been proposed from time to time. One of the earliest of these, described by A. Werner in 1905, is notable because it divides each of the shorter periods into two parts, one at either end of the table over the elements in the longer periods that they most resemble. The multiple tie lines connecting the periods in the Bayley-type table are thereby dispensed with. This class of table, too, can be greatly simplified by removing the lanthanon and actinon elements to a separate area. By the mid-20th century this version of the table (Figure 10) had become the most commonly used.

PREDICTIVE VALUE OF THE PERIODIC LAW

Discovery of new elements. The great value of the periodic law was made evident by Mendeleyev's success in 1871 in finding that the properties of 17 elements could be correlated with those of other elements by moving the 17 to new positions from those indicated by their atomic weights. This change indicated that there were small errors in the previously accepted atomic weights of several of the elements and large errors for several others, for which wrong multiples of the combining weights had been used as atomic weights (the combining weight being that weight of an element that combines with a given weight of a standard). Mendeleyev was also able to predict the existence, and many of the properties, of the then undiscovered elements ekaboron, ekaaluminum, and

Group 0	I		II		III		IV		V		VI		VII		VIII
	a	b	a	b	a	b	a	b	a	b	a	b	a	b	
	H 1														
He 2	Li 3		Be 4		B 5			C 6		N 7		O 8		F 9	
Ne 10	Na 11		Mg 12		Al 13			Si 14		P 15		S 16		Cl 17	
Ar 18	K 19	Cu 29	Ca 20	Zn 30	Sc 21	Ga 31	Ti 22	Ge 32	V 23	As 33	Cr 24	Se 34	Mn 25	Br 35	Fe 26, Co 27, Ni 28
Kr 36	Rb 37	Ag 47	Sr 38	Cd 48	Y 39	In 49	Zr 40	Sn 50	Nb 41	Sb 51	Mo 42	Te 52	Tc 43	I 53	Ru 44, Rh 45, Pd 46
Xe 54	Cs 55	Au 79	Ba 56	Hg 80	57–71*	Tl 81	Hf 72	Pb 82	Ta 73	Bi 83	W 74	Po 84	Re 75	At 85	Os 76, Ir 77, Pt 78
Rn 86	Fr 87		Ra 88		Ac 89		Th 90		Pa 91		U 92		Np 93		Pu 94, Am 95, Cm 96

*Lanthanum and the lanthanons

Figure 7: Short-period form of periodic system of elements.

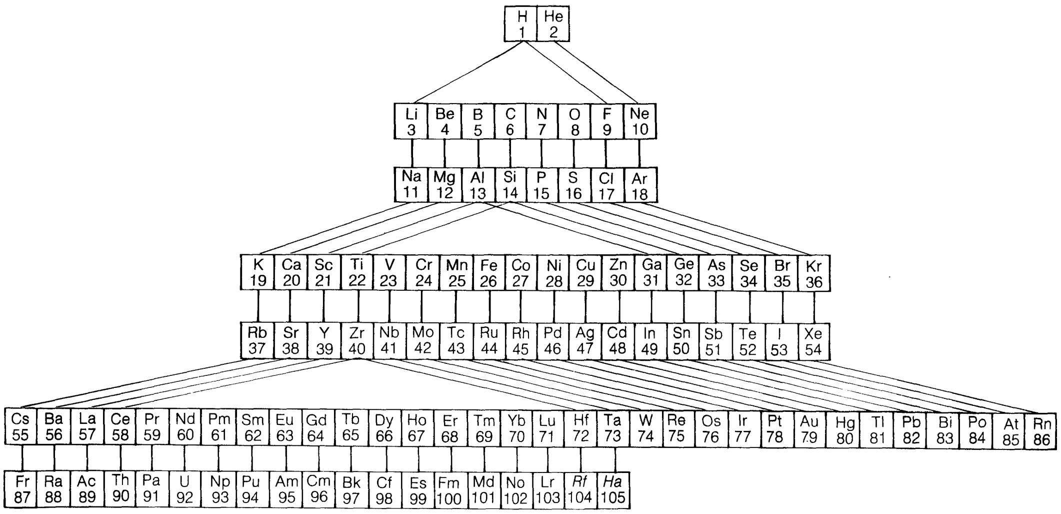


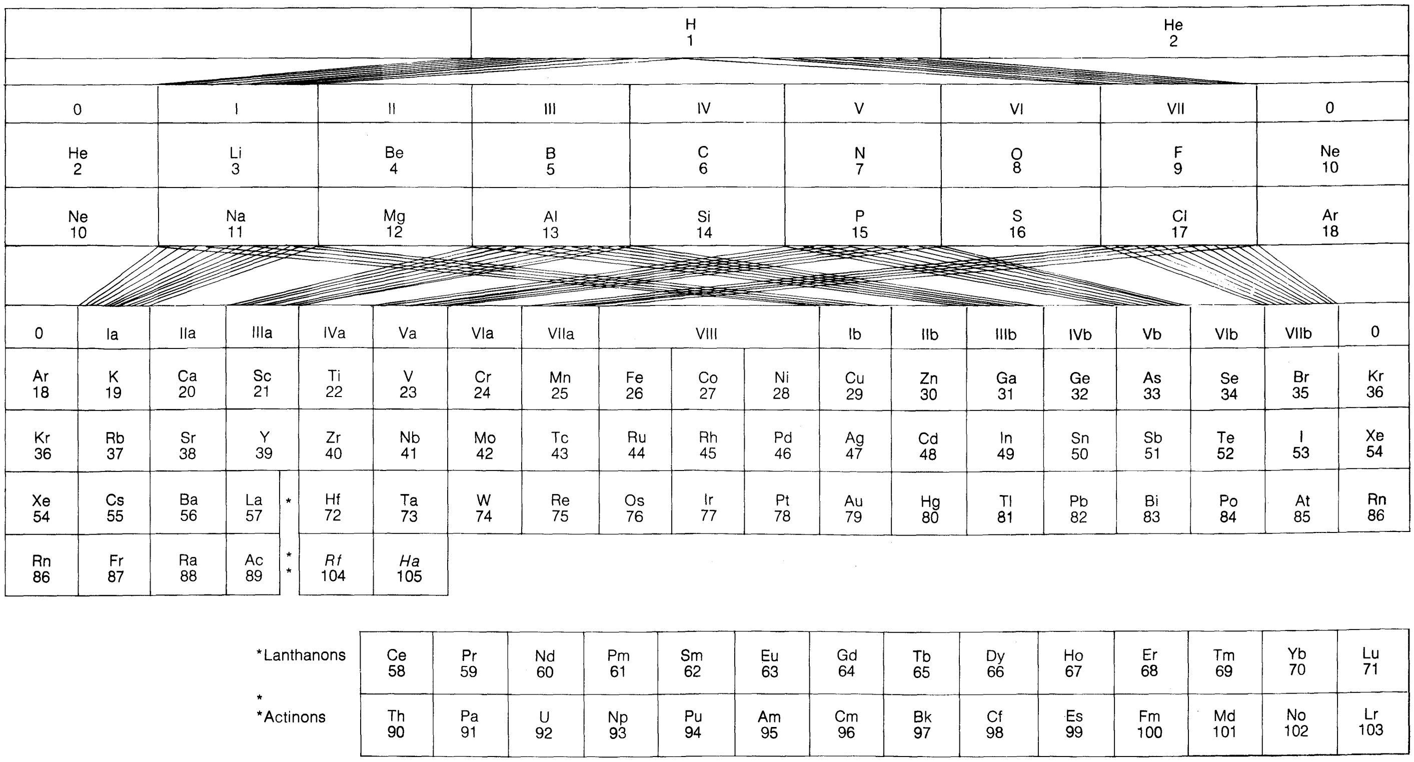
Figure 8: Periodic system of elements with periods demarcated by noble gases.

ekasilicon, now identified with the elements scandium, gallium, and germanium, respectively. Similarly, after the discovery of helium and argon, the periodic law permitted the prediction of the existence of neon, krypton, xenon, and radon. Moreover, Bohr pointed out that the missing element 72 would be expected, from its position in the periodic system, to be similar to zirconium in its properties rather than to the rare earths; this observation led G. de Hevesy and D. Coster in 1922 to examine zirconium ores and to discover the unknown element, which they named hafnium.

In spite of the corrections made by the redetermination of atomic weights, some of the elements in the Mendeleyev and Lothar Meyer periodic tables of 1871 were still required by their properties to be put in positions somewhat out of the order of atomic weights. Thus, in the pairs argon and potassium, cobalt and nickel, and tellurium and iodine, the first element has the greater atomic weight but the earlier position in the periodic system. The eventual solution to this difficulty was found only when the structure of the atom was better understood.

Significance of the atomic number

About 1910, Sir Ernest Rutherford's experiments on the scattering of alpha particles by the nuclei of heavy atoms led to the determination of the nuclear electrical charge. The ratio of the nuclear charge to that of the electron was noted to be roughly one-half the atomic weight. In 1911 A. van den Broek suggested that this quantity, the atomic number, might be identified with the ordinal number of the element in the periodic system (following the lead of Newlands, it had become customary to number the elements according to their position in the table). This suggestion was brilliantly confirmed in 1913 by H.G.J. Moseley's measurements of the wavelengths of the characteristic X-ray spectral lines of many elements, which showed that the wavelengths did indeed depend in a regular way upon the atomic numbers—identical with the ordinal numbers of the elements in the table. There is no longer any uncertainty about the position of any element in the ordered series of the periodic system. All of the elements from 1 to 107 are now known, and it is to be expected that new elements beyond 107 will be made by nuclear reactions.



	H 1	He 2

0	I	II	III	IV	V	VI	VII	0
He 2	Li 3	Be 4	B 5	C 6	N 7	O 8	F 9	Ne 10
Ne 10	Na 11	Mg 12	Al 13	Si 14	P 15	S 16	Cl 17	Ar 18

0	Ia	IIa	IIIa	IVa	Va	VIa	VIIa	VIII			Ib	IIb	IIIb	IVb	Vb	VIb	VIIb	0
Ar 18	K 19	Ca 20	Sc 21	Ti 22	V 23	Cr 24	Mn 25	Fe 26	Co 27	Ni 28	Cu 29	Zn 30	Ga 31	Ge 32	As 33	Se 34	Br 35	Kr 36
Kr 36	Rb 37	Sr 38	Y 39	Zr 40	Nb 41	Mo 42	Tc 43	Ru 44	Rh 45	Pd 46	Ag 47	Cd 48	In 49	Sn 50	Sb 51	Te 52	I 53	Xe 54
Xe 54	Cs 55	Ba 56	La 57 *	Hf 72	Ta 73	W 74	Re 75	Os 76	Ir 77	Pt 78	Au 79	Hg 80	Tl 81	Pb 82	Bi 83	Po 84	At 85	Rn 86
Rn 86	Fr 87	Ra 88	Ac 89 **	Rf 104	Ha 105													

*Lanthanons	Ce 58	Pr 59	Nd 60	Pm 61	Sm 62	Eu 63	Gd 64	Tb 65	Dy 66	Ho 67	Er 68	Tm 69	Yb 70	Lu 71
**Actinons	Th 90	Pa 91	U 92	Np 93	Pu 94	Am 95	Cm 96	Bk 97	Cf 98	Es 99	Fm 100	Md 101	No 102	Lr 103

Figure 9: Long-period form of periodic system of elements.

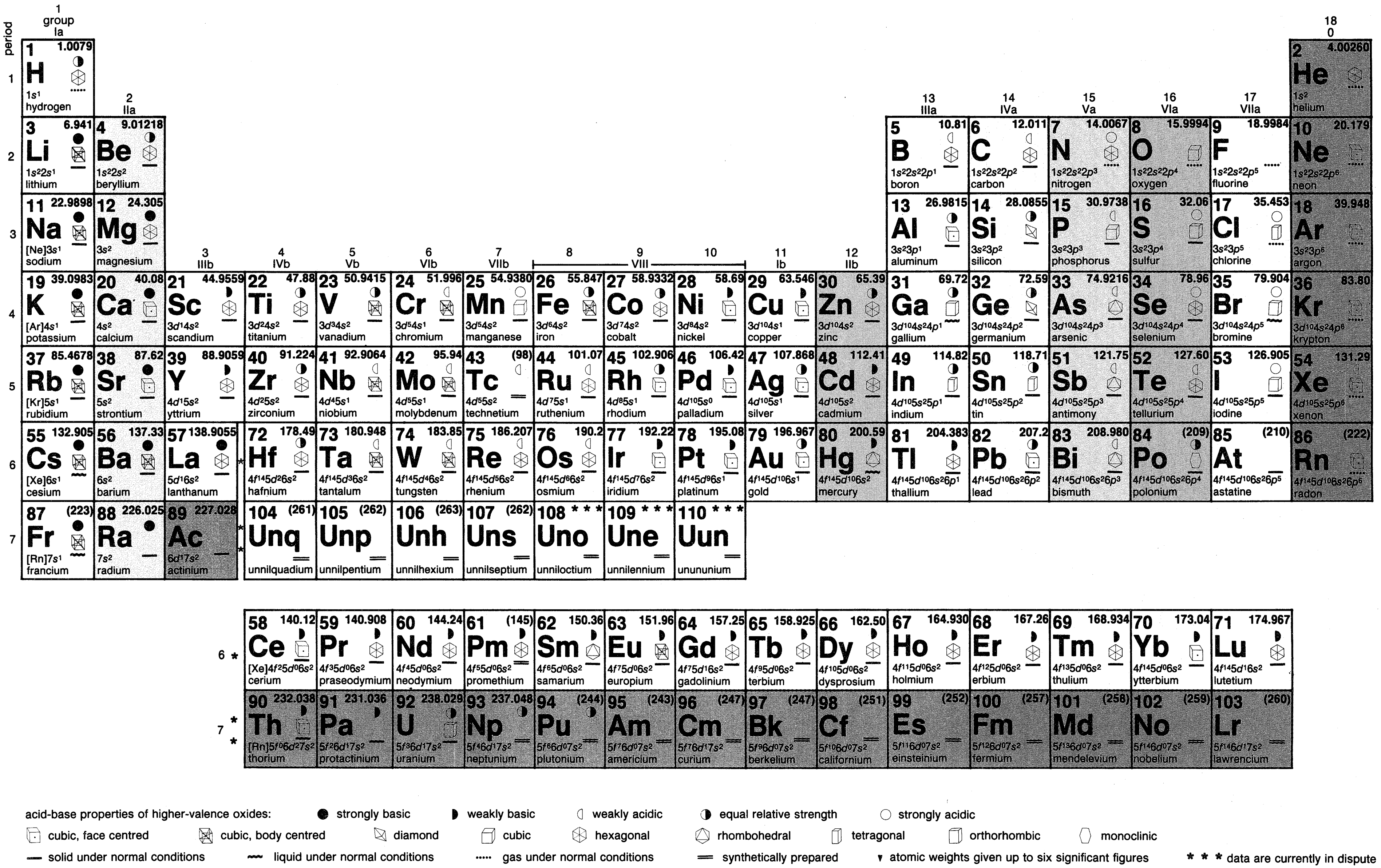


Figure 10: Modern version of the periodic table of the elements.

That the exact atomic weight of an element is of small significance for its position in the periodic system is shown by the existence of isotopes of every element, atoms with the same atomic number but different atomic weights. The chemical properties of the isotopes of an element are essentially the same, and all of the isotopes of an element occupy the same place in the periodic system in spite of their differences in atomic weight.

Elucidation of the periodic law. Detailed understanding of the periodic system has developed along with the quantum theory of spectra and the electronic structure of atoms, beginning with the work of Bohr in 1913. Important forward steps were the formulation of the general rules of the old quantum theory by W. Wilson and A. Sommerfeld in 1916, the discovery of the exclusion principle by W. Pauli in 1925 (see ATOMS), the discovery of the spin of the electron by G.E. Uhlenbeck and S. Goudsmit in 1925, and the development of quantum mechanics (see MECHANICS: *Quantum mechanics*) by W. Heisenberg and E. Schrödinger during the same year. The development of the electronic theory of valence and molecular structure (see MOLECULES), beginning with the postulate of the shared electron pair by G.N. Lewis in 1916, also played a very important part in explaining the periodic law (see MOLECULES and CHEMICAL REACTIONS).

The long form of the periodic table

PERIODS

The long-period form of the periodic system of the elements, presented here in two forms, Figures 9 and 10, contains all of the chemical elements that have been discovered or made; they are arranged, in the order of their atomic numbers, in seven horizontal periods, with the lanthanons (lanthanum, 57, to lutetium, 71) and the actinons (actinium, 89, to lawrencium, 103) indicated separately below (unless otherwise stated, Figure 9 will be used as reference). The periods are of varying lengths. First there is the hydrogen period, consisting of the two elements hydrogen, 1, and helium, 2. Then there are two periods of eight elements each: the first short period, from lithium, 3, to neon, 10; and the second short period, from sodium, 11, to argon, 18. There follow two periods of 18 elements each: the first long period, from potassium 19, to krypton, 36; and the second long period, from rubidium, 37, to xenon, 54. The very long period of 32 elements, from cesium, 55, to radon, 86, is condensed into 18 columns by the omission of the lanthanons (which are indicated separately below), permitting the remaining 18 elements, which are closely similar in their properties to corresponding elements of the first and second long periods, to be placed directly below these elements. The last period is incomplete. The properties of these elements indicate that this period is a second very long period of 32 elements, which would be completed at element 118.

GROUPS

The six noble gases, helium, neon, argon, krypton, xenon, and radon, occur at the ends of the six completed periods and constitute the 0 group of the periodic system. (It is customary to refer to horizontal series of elements in the table as periods and vertical series as groups.) The seven elements lithium to fluorine and the seven corresponding elements sodium to chlorine are placed, in Figure 9, in the seven groups, I, II, III, IV, V, VI, and VII, respectively. The 17 elements of the fourth period, from potassium, 19, to bromine, 35, are distinct in their properties and might well be considered to constitute 17 subgroups of the periodic system. Largely for historical reasons (the use of the short-period form of the periodic system, Figure 7), however, it is customary to describe them as constituting 15 subgroups, the three consecutive elements iron, cobalt, and nickel and their congeners in the succeeding periods being placed in a single group, VIII. The elements potassium, 19, to manganese, 25, are assigned to subgroups Ia, IIa, IIIa, IVa, Va, VIa, and VIIa, respectively; and the elements copper, 29, to bromine, 35, are assigned to subgroups Ib, IIb, IIIb, IVb, Vb, VIb, and VIIb. (In some modern versions of the periodic system, the members of subgroups a and b in Groups III and above are the reverse of the arrangement shown here. There is no one standard form of the system.)

Subgroups

The first group, the alkali metals, thereby includes, in addition to lithium and sodium, the metals from potassium down the table to francium, but not the much less similar metals of subgroup Ib (copper, etc.). Also the second group, the alkaline-earth metals, is considered to include beryllium, magnesium, calcium, strontium, barium, and radium but not the elements of subgroup IIb. The properties of the elements do not indicate unambiguously whether the boron group might be considered better to include the subgroup IIIa or the subgroup IIIb, although the former is often so included. The other four groups are as follows: the carbon group, IV, consists of carbon, silicon, germanium, tin, and lead; the nitrogen group, V, includes nitrogen, phosphorus, arsenic, antimony, and bismuth; the oxygen group, VI, includes oxygen, sulfur, selenium, tellurium, and polonium; and the halogen group, VII, includes fluorine, chlorine, bromine, iodine, and astatine.

Although hydrogen is included in Group I in some versions of the table, it is not closely similar to either the alkali metals or the halogens in its chemical properties. It is, however, assigned the oxidation number +1 in compounds such as hydrogen fluoride, HF, and −1 in compounds such as lithium hydride, LiH; and it may hence be considered as being similar to a Group Ia element and to a Group VII element, respectively, in compounds of these two types, taking the place first of Li and then of F in lithium fluoride, LiF. Hydrogen is, in fact, the most individualistic of the elements: no other element resembles it in the way that sodium resembles lithium, chlorine resembles fluorine, and neon resembles helium. It is a unique element, the only element that cannot conveniently be considered a member of a group.

A number of the elements of each long period are called the transition metals. These are usually taken to be scandium, 21, to zinc, 30 (the iron-group transition metals); yttrium, 39, to cadmium, 48 (the palladium-group transition metals); and hafnium, 72, to mercury, 80 (the platinum-group transition metals). By this definition, the transition metals include Groups IIIa to VIIa, VIII, and Ib and IIb (Figure 9) in the long periods.

Periodic properties

The periodicity in properties of the elements arranged in order of atomic number is strikingly shown by the consideration of the physical state of the elementary substances and such related properties as the melting point, density, and hardness. The elements of Group 0 are gases that are difficult to condense. The alkali metals, in Group I, are soft, metallic solids with low melting points. The alkaline-earth metals, in Group II, are harder and have higher melting points than the adjacent alkali metals. The hardness and melting point continue to increase through Groups III and IV and then decrease through Groups V, VI, and VII. The elements of the long periods show a gradual increase in hardness and melting point from the beginning alkali metals to near the centre of the period and then at Group VIb an irregular decrease to the halogens and noble gases.

The valence of the elements (that is, the number of bonds formed with a standard element) is closely correlated with position in the periodic table, the elements in the main groups having maximum positive valence, or oxidation number, equal to the group number, and maximum negative valence equal to the difference between eight and the group number.

The general chemical properties described as metallic or base forming, metalloid or amphoteric, and nonmetallic, or acid forming, are correlated with the periodic table in a simple way: the most metallic elements are to the left and to the bottom of the periodic table, and the most nonmetallic elements are to the right and to the top (ignoring the noble gases). The metalloids are adjacent to a diagonal line from boron to polonium. A closely related property is electronegativity, the tendency of atoms to retain their electrons and to attract additional electrons. The degree of electronegativity of an element is shown by ionization potential, electron affinity, oxidation-reduction potential, the energy of formation of chemical bonds, and other properties. It is shown to depend upon the element's position in the periodic table in the same way that nonmetallic character does, fluorine being the most electronegative element and cesium (or francium) the least electronegative (most electropositive) element.

Atomic sizes

The sizes of atoms of elements vary regularly throughout the periodic system. Thus, the effective bonding radius (or one-half the distance between adjacent atoms) in the elementary substances in their crystalline or molecular forms decreases through the first short period from 1.52 Å for lithium to 0.73 Å for fluorine; at the beginning of the second period, the bonding radius rises abruptly to 1.86 Å for sodium and gradually decreases to 0.99 Å for chlorine. The behaviour through the long periods is more complex: the bonding radius decreases gradually from 2.31 Å for potassium to a minimum of 1.25 Å for cobalt and nickel, then rises slightly, and finally falls to 1.14 Å for bromine. The sizes of atoms are of importance in the determination of coordination number (that is, the number of groups attached to the central atom in a compound) and hence in the composition of compounds. The increase in atomic size from the upper right corner of the periodic table to the lower left corner is reflected in the formulas of the oxygen acids of the elements in their highest states of oxidation (see Table 9). The smallest atoms group only three oxygen atoms about themselves; the next larger atoms, which coordinate a tetrahedron of four oxygen atoms, are in a diagonal belt; and the still larger atoms, which form octahedral oxygen complexes (stannic acid, antimonic acid, telluric acid, paraperiodic acid), lie below and to the left of this belt. Only the chemical and physical properties of the elements are determined by the extranuclear electronic structure; these properties show the periodicity described in the periodic law. The properties of the atomic nuclei themselves, such as the magnitude of the packing fraction and the power of entering into nuclear reactions, are, although dependent upon the atomic number, not dependent in the same periodic way.

Table 9: Formulas of Oxygen Acids of Elements in Their Highest Oxidation State

$H_2Be(OH)_4$	H_3BO_3	H_2CO_3	HNO_3		
	$HAl(OH)_4$	H_4SiO_4	H_3PO_4	H_2SO_4	$HClO_4$
		H_4GeO_4	H_3AsO_4	H_2SeO_4	
		$H_2Sn(OH)_6$	$HSb(OH)_6$	H_6TeO_6 and H_2TeO_4	HIO_4 and H_5IO_6

The basis of the periodic system

ELECTRONIC STRUCTURE

The noble gases—helium, neon, argon, krypton, xenon, and radon—have the striking chemical property of forming few chemical compounds. This property would depend upon their possessing especially stable electronic structures (that is, structures so firmly knit that they would not yield to accommodate ordinary chemical bonds). During the development of modern atomic physics and the theory of quantum mechanics, a precise and detailed understanding was obtained of the electronic structure of the noble gases and other atoms that explains the periodic law in a thoroughly satisfactory manner.

Quantum numbers

The Pauli exclusion principle states that no more than two electrons can occupy the same orbit—or, in quantum-mechanical language, orbital—in an atom and that two electrons in the same orbital must be paired (that is, must have their spins opposed). The orbitals in an atom may be described by a principal quantum number, n, which may assume the values 1, 2, 3, . . . , and by an azimuthal quantum number, l, which may assume the values 0, 1, 2, . . . , $n-1$. There are $2l+1$ distinct orbitals for each set of values of n and l. The most stable orbitals, which bring the electron closest to the nucleus, are those with the smallest values of n and l. The electrons that occupy the orbital with $n=1$ (and $l=0$) are said to be in the *K* shell of electrons; the *L*, *M*, *N*, . . . shells correspond respectively to $n=2$, 3, 4, ... Each shell except the *K* shell is divided into subshells corresponding to the values 0, 1, 2, 3, . . . of the orbital quantum number l; these subshells are called the *s*, *p*, *d*, *f*, . . . subshells, and they can accommodate a maximum of 2, 6, 10, 14, . . . electrons. (There is no special significance to the letter designations of the quantum numbers or of the shells and subshells.)

The approximate order of stability of the successive subshells in an atom is indicated in the chart below. The number of electrons in the atoms of the elements increases with increasing atomic number, and the added electrons go, of necessity, into successively less stable shells. The most stable shell, the *K* shell, is completed with helium, which has two electrons. The *L* shell is then completely filled at neon, with atomic number 10. The atoms of the heavier noble gases do not, however, have a completed outer shell but instead have *s* and *p* subshells only. The outer shell of eight electrons is called traditionally an octet. The *d* subshells and *f* subshells subsequently are also filled with electrons after the initially less stable orbitals are occupied, an inversion of stability having occurred with increasing atomic number.

The electron occupancy of the shells in the noble gas atoms is as follows:

Electrons in	*K*	*L*	*M*	*N*	*O*	*P*	*Q*	shell
He 2:	2							
Ne 10:	2	8						
Ar 18:	2	8	8					
Kr 36:	2	8	18	8				
Xe 54:	2	8	18	18	8			
Rn 86:	2	8	18	32	18	8		
118:	2	8	18	32	32	18	8	

The numbers 2, 8, 18, and 32 correspond to filling the *s*; *s* and *p*; *s*, *p*, and *d*; and *s*, *p*, *d*, and *f* subshells, respectively.

Periods and electron shells

The first period of the periodic table is complete at helium, when the *K* shell is filled with two electrons. The first and second short periods represent the filling of the 2*s* and 2*p* subshells (completing the *L* shell at neon) and the 3*s* and 3*p* subshells (at argon), leaving the *M* shell incomplete. The first long period begins with the introduction of electrons into the 4*s* orbital. Then, at scandium, the five 3*d* orbitals of the inner *M* shell begin to be occupied. It is the successive occupancy of these five 3*d* orbitals by their complement of ten electrons that characterizes the ten elements of the iron-group transition series. At krypton the *M* shell is complete and there is an octet in the *N* shell. The second long period, of 18 elements, similarly represents the completion of an outer octet and the next inner subshell of ten 4*d* electrons.

The very long period of 32 elements results from the completion of the 4*f* subshell of 14 electrons, the 5*d* subshell of 10 electrons, and the 6*s*, 6*p* octet. The filling of the 4*f* orbitals corresponds to the sequence of 14 lanthanons, and that of the 5*d* orbitals to the 10 platinum-group transition metals.

The next period, which is incomplete, involves the 5*f* subshell of 14 electrons, the 6*d* subshell of 10 electrons, and the 7*s*, 7*p* octet. The filling of the 5*f* orbitals corresponds to the actinons, the elements beginning with thorium, atomic number 90.

The successive periods of the system hence correspond to the introduction of electrons into the following orbitals:

The very short period, H, He.	2 elements. 1*s*.
The first short period, Li to Ne.	8 elements. 2*s*, 2*p*.
The second short period, Na to Ar.	8 elements. 3*s*, 3*p*.
The first long period, K to Kr.	18 elements. 3*d*, 4*s*, 4*p*.
The second long period, Rb to Xe.	18 elements. 4*d*, 5*s*, 5*p*.
The very long period, Cs to Rn.	32 elements. 4*f*, 5*d*, 6*s*, 6*p*.
The incomplete period, Fr to 118.	32 elements. 5*f*, 6*d*, 7*s*, 7*p*.

There are advantages to replacing the *K*, *L*, *M*, . . . shells by a different grouping of the subshells, in which those with nearly the same energy are grouped together, in close correlation with the periodic system. The new set of shells is the following:

The helium shell.	2 electrons. 1*s*.
The neon shell.	8 electrons. 2*s*, 2*p*.
The argon shell.	8 electrons. 3*s*, 3*p*.
The krypton shell.	18 electrons. 4*s*, 4*p*, 3*d*.
The xenon shell.	18 electrons. 5*s*, 5*p*, 4*d*.
The radon shell.	32 electrons. 6*s*, 6*p*, 5*d*, 4*f*.
The eka-radon shell.	32 electrons. 7*s*, 7*p*, 6*d*, 5*f*.

PERIODICITY IN PROPERTIES OF THE ELEMENTS

Properties and electronic structure

The periodicity of properties of the elements is caused by the periodicity in electronic structure. The noble gases are chemically unreactive, or nearly so, because their electronic structures are stable—their atoms hold their quota of electrons strongly, have no affinity for more electrons, and have little tendency to share electrons with other atoms. An element close to a noble gas in the periodic system, on the other hand, is reactive chemically because of the possibility of assuming the stable electronic configuration of the noble gas, by losing one or more electrons to another atom, by gaining one or more electrons from another atom, or by sharing electrons. The alkali metals, in Group I, can assume the noble-gas configuration by losing one electron, which is loosely held in the outermost (valence) shell, to another element with greater electron affinity, thus producing the stable singly charged positive ions. Similarly the alkaline-earth metals can form doubly charged positive ions with the noble-gas electronic configuration by losing the two loosely held electrons of the valence shell; the positive ionic valences of the elements of the first groups are hence equal to the group numbers. The elements just preceding the noble gases can form negative ions with the noble-gas configuration by gaining electrons; the negative ionic valences of these elements are equal to the difference between eight and their group numbers. The covalence (or number of shared electron pairs) of an atom is determined by its electron number and the stable orbitals available to it. An atom such as fluorine, with seven electrons in its outer shell, can combine with a similar atom by sharing a pair of electrons with it; each atom thus achieves the noble-gas configuration by having three unshared pairs and one shared electron pair in its valence shell.

The properties of elements in the same group of the periodic system are, although similar, not identical. The trend in properties from the lighter to the heavier elements may be attributed to changes in the strength of binding of the outer electrons and especially to the increasing size of the atoms. (L.C.P.)

Other chemical and physical classifications

Certain methods of classifying elements on the basis of chemical properties are not strictly related to the groups in which the elements appear. Such classification schemes illustrate the fact that useful horizontal as well as vertical relationships exist in the periodic table. Thus, the transition elements, either as a whole or as three horizontal series, are often considered together when chemical properties are discussed. The transition elements in each horizontal series exhibit much less variation in atomic size than do the elements in other parts of the same periods, leading to a similarity in chemical and physical properties. The lanthanide and actinide elements exhibit an even greater similarity for the same reason. The metallic elements in Groups Ia and IIa are often classed together because they are markedly more reactive than the other metallic elements. At the other extreme, elements of the platinum group—including ruthenium, rhodium, palladium, osmium, iridium, and platinum—are chemically inert, as are silver and gold; these elements are collectively designated the noble metals because they do not readily enter into combination with other elements.

Solid, liquid, and gaseous elements

Of all the 107 known elements, 11 are gaseous, 2 are liquid, and the remainder are solids under ordinary conditions. Except for hydrogen and mercury, the gaseous and liquid elements occur in the right-hand part of the periodic table, the region associated with the nonmetallic elements.

The physical characteristics of the elements provide convenient means of identification. The melting points of the various elements range from −272° C (for helium) to greater than 3,500° C (for carbon in the form of diamond). Properties such as boiling points, electrical conductivity, and thermal conductivity also can be used for identification because they are unique for each element. Perhaps the single most useful characteristic for identifying an element is its pattern of light absorption, or emission, which is called a spectrum. An element exhibits its own characteristic spectrum whether it exists in the free state, in a mixture, or in chemical combination with other elements. Since the intensity of the spectrum is dependent upon the amount of the element contained in the sample, the spectrum also can be used as a means for quantitative analysis of the elements. There are several chemical methods for estimating the percentage of an element present in a sample; these, however, require a detailed knowledge of the chemistry of the element in question (see ANALYSIS AND MEASUREMENT).

Man-made elements

All naturally occurring elements with atomic numbers greater than 84 are radioactive; these constitute the three families of radioactive elements (*i.e.,* the uranium, thorium, and actinide series), the elements of which are converted from one to another by radioactive processes. In addition, several naturally occurring isotopes of the lighter elements are radioactive. The atomic nuclei of all radioactive elements are unstable and emit highly energetic particles. In the process, the number of protons in the nucleus changes, and the atom is transformed into one of a different element. The half-life of a radioactive isotope is the time required for half of any amount of the isotope to disintegrate by radioactive decay. The common modes of decay of radioactive isotopes are loss of beta or alpha particles or the capture of an electron. The loss of a beta particle, or electron, from the nucleus increases the atomic number by one unit; the loss of an alpha particle, or helium nucleus (two protons and two neutrons), decreases the atomic number by two units; and the process of electron capture, in which an electron from an inner shell is drawn into the nucleus, corresponds to a decrease of atomic number by one unit. Eleven radioactive synthetic, or man-made, elements (see below *Transuranium elements*) have not as yet been found in nature and the presence of two nontransuranium elements has not been verified absolutely. Although the remaining elements generally are not considered to be radioactive, some do have radioactive isotopes that exist naturally in very small concentrations, and more than 1,000 radioactive isotopes of these elements have been prepared in the laboratory. (J.J.L./Ed.)

THE ELEMENT GROUPS

Hydrogen, the introductory element to the entire periodic classification, has unusual properties. In some of its reactions hydrogen resembles the alkali metals, but in others it appears closely akin to the halogens. It is the only element that cannot be considered a distinct member of any particular group and is therefore often, as here, treated by itself. The balance of the section treats the main groups of the periodic table, the transition elements, and the related inner transition series—the lanthanides and the actinides. The transuranium elements beyond those usually included among the actinides, the so-called transactinides, are also discussed in brief and the possible existence of superheavy elements (*i.e.,* elements with atomic numbers of 110 or higher) is considered. (Ed.)

Hydrogen

Although hydrogen is the most abundant element in the Galaxy (ten times as abundant as helium, the next most widely occurring element), it makes up only about 0.14 percent of the Earth's crust by weight. It occurs, however, in vast quantities as part of the water in oceans, ice packs, rivers, lakes, and the atmosphere. As part of innumerable carbon compounds, hydrogen is present in all animal and vegetable tissue and in petroleum. Even though it is often said that there are more known compounds of carbon than of any other element, the fact is that since hydrogen is contained in almost all carbon compounds and also forms a multitude of compounds with all other elements (except some of the noble gases), it is possible that hydrogen compounds are more numerous. Hydrogen is an odourless, tasteless, colourless gas; the lightest and simplest chemical element. A hydrogen atom, represented by the symbol H, has, for its nucleus, a single proton (*i.e.,* a particle of matter with a positive charge equal to the opposite, negative charge of an electron) surrounded by one orbiting electron. Under ordinary conditions, hydrogen gas is a loose aggregation of hydrogen molecules, each consisting of a pair of atoms, a diatomic molecule, H_2. The earliest known important chemical property of hydrogen is that it burns with oxygen to form water, H_2O; indeed, the name hydrogen is derived from Greek words meaning "maker of water."

Elementary hydrogen finds its principal industrial application in the manufacture of ammonia (a compound of hydrogen and nitrogen, NH_3) and in the hydrogenation of carbon monoxide and organic compounds.

Hydrogen has three known isotopes. The mass numbers of hydrogen's isotopes are 1, 2, and 3, the most abundant being the mass 1 isotope generally called hydrogen (symbol H, or ^1H) but also known as protium. The mass 2 isotope, which has a nucleus of one proton and one neutron and has been named deuterium, or heavy hydrogen (symbol D, or ^2H), constitutes 0.0156 percent of the ordinary mixture of hydrogen. Tritium (symbol T, or ^3H), with one proton and two neutrons in each nucleus, is the mass 3 isotope and constitutes about 10^{-15} to 10^{-16} percent of hydrogen. The practice of giving distinct names to the hydrogen isotopes is justified by the fact that there are significant differences in their properties.

Early experimentation with hydrogen

Paracelsus, physician and alchemist, in the 16th century unknowingly experimented with hydrogen when he found that an inflammable gas was evolved when a metal was dissolved in acid. The gas, however, was confused with other inflammable gases, such as hydrocarbons and carbon monoxide. In 1766 Henry Cavendish, English chemist and physicist, showed that hydrogen, then called inflammable air, phlogiston, or the inflammable principle, was distinct from other combustible gases because of its density and the amount of it that evolved from a given amount of acid and metal. In 1781 Cavendish confirmed previous observations that water was formed when hydrogen was burned, and Antoine-Laurent Lavoisier, the father of modern chemistry, coined the French word *hydrogène* from which the English form is derived. In 1929 Karl Friedrich Bonhoeffer, a German physical chemist, and Paul Harteck, an Austrian chemist, on the basis of earlier theoretical work, showed that ordinary hydrogen is a mixture of two kinds of molecules, *ortho*-hydrogen and *para*-hydrogen. Because of the simple structure of hydrogen, its properties can be theoretically calculated relatively easily. Hence hydrogen is often used as a theoretical model for more complex atoms, and the results are applied qualitatively to other atoms.

Table 10: Some Properties of Normal Hydrogen and Deuterium

	normal hydrogen	deuterium
Atomic hydrogen		
Atomic number	1	1
Atomic weight	1.0080	2.0141
Ionization potential	13.595 eV	13.600 eV
Electron affinity	0.754 eV	—
Nuclear spin	½	1
Nuclear magnetic moment (nuclear magnetons)	2.7927	0.8574
Nuclear quadrupole moment	0	2.77×10^{-27} cm^2
Electronegativity (Pauling)	2.1	~2.1
Molecular hydrogen		
Bond distance	0.7416 Å	0.7416 Å
Dissociation energy (25° C)	104.19 kcal/mole	105.97 kcal/mole
Ionization potential	15.427 eV	15.457 eV
Density of solid	0.08671 g/cm^3	0.1967 g/cm^3
Melting point	−259.20° C	−254.43° C
Heat of fusion	28 cal/mole	47 cal/mole
Density of liquid	0.07099 (−252.78°)	0.1630 (−249.75°)
Boiling point	−252.77° C	−249.49° C
Heat of vaporization	216 cal/mole	293 cal/mole
Critical temperature	−240.0° C	−234.8° C
Critical pressure	13.0 atm	16.4 atm
Critical density	0.0310 g/cm^3	0.0668 g/cm^3
Heat of combustion to H_2O(g)	−57.796 kcal/mole	−59.564 kcal/mole

PHYSICAL AND CHEMICAL PROPERTIES

Table 10 lists the important properties of molecular hydrogen, H_2. The extremely low melting and boiling points result from weak forces of attraction between the molecules. The existence of these weak intermolecular forces is also revealed by the fact that, when hydrogen gas expands from high to low pressure at room temperature, its temperature rises, whereas the temperature of most other gases falls. According to thermodynamic principles, this implies that repulsive forces exceed attractive forces between hydrogen molecules at room temperature, otherwise the expansion would cool hydrogen. In fact, at −68.6° C attractive forces predominate, and hydrogen, therefore, cools upon being allowed to expand below that temperature. The cooling effect becomes so pronounced at temperatures below that of liquid nitrogen (−196° C) that the effect is utilized to achieve the liquefaction temperature of hydrogen gas itself.

Hydrogen is transparent to visible light, to infrared light, and to ultraviolet light to wavelengths below 1800 Å. Because its molecular weight is lower than that of any other gas, its molecules have a velocity higher than those of any other gas at a given temperature and it diffuses faster than any other gas. Consequently, kinetic energy is distributed faster through hydrogen than through any other gas; it has, for example, the greatest heat conductivity.

The hydrogen molecule

A molecule of hydrogen is the simplest possible molecule. It consists of two protons and two electrons held together by electrostatic forces. Like atomic hydrogen, the assemblage can exist in a number of energy levels.

Ortho-hydrogen and para-hydrogen. Two types of molecular hydrogen (*ortho* and *para*) are known. These differ in the magnetic interactions of the protons due to the spinning motions of the protons. In *ortho*-hydrogen, the spins of both protons are aligned in the same direction—that is, they are parallel. In *para*-hydrogen, the spins are aligned in opposite directions and are therefore antiparallel. The relationship of spin alignments determines the magnetic properties of the atoms. Normally, transformations of one type into the other (*i.e.,* conversions between *ortho* and *para* molecules) do not occur and *ortho*-hydrogen and *para*-hydrogen can be regarded as two distinct modifications of hydrogen. The two forms may, however, interconvert under certain conditions. Equilibrium between the two forms can be established in several ways. One of these is by the introduction of catalysts (such as activated charcoal or various paramagnetic substances); another method is to apply an electrical discharge to the gas or to heat it to a high temperature.

The concentration of *para*-hydrogen in a mixture that has achieved equilibrium between the two forms depends on the temperature as shown by the following figures:

−253.1° C	99.82%	−153.1° C	32.87%
−223.1° C	76.89%	0° C	25.13%
−193.1° C	48.39%	200° C	25.00%

Isolation of *para*-hydrogen

Essentially pure *para*-hydrogen can be produced by bringing the mixture into contact with charcoal at the temperature of liquid hydrogen; this converts all the *ortho* into *para*. The *ortho*-hydrogen, on the other hand, cannot be prepared directly from the mixture because the concentration of *para*-hydrogen is never less than 25 percent.

The two forms of hydrogen have slightly different physical properties. The melting point of *para*-hydrogen is 0.10° lower than that of a 3:1 mixture of *ortho*-hydrogen and *para*-hydrogen. At −252.77° C the pressure exerted by the vapour over liquid *para*-hydrogen is 1.035 atmospheres (one atmosphere is the pressure of the atmosphere at sea level under standard conditions, equal to about 14.69 pounds per square inch), compared with 1.000 atmosphere for the vapour pressure of the 3:1 *ortho–para* mixture. As a result of the different vapour pressures of *para*-hydrogen and *ortho*-hydrogen, these forms of hydrogen can be separated by low-temperature gas chromatography, an analytical process that separates different atomic and molecular species on the basis of their differing volatilities.

Reactivity of hydrogen. One molecule of hydrogen dissociates into two atoms ($H_2 \rightarrow 2H$) when an energy equal to or greater than the dissociation energy (*i.e.,* the amount of energy required to break the bond that holds together the atoms in the molecule) is supplied. The dissociation energy of molecular hydrogen is 104,000 calories per mole—written 104 kcal/mole (mole: the molecular weight expressed in grams, which is two grams in the case of hydrogen). Sufficient energy is obtained, for example, when the gas is brought into contact with a white-hot tungsten filament or when an electric discharge is established in the gas. If atomic hydrogen is generated in a system at low pressure, the atoms will have a significant lifetime—*e.g.,* 0.3 second at a pressure of 0.5 millimetre of mercury. Atomic hydrogen is very reactive. It combines with most elements to form hydrides (*e.g.,* sodium hydride, NaH), and it reduces metallic oxides, a reaction that produces the metal in its elemental state. The surfaces of metals that do not combine with hydrogen to form stable hydrides (*e.g.,* platinum) catalyze the recombination of hydrogen atoms to form hydrogen molecules and are thereby heated to incandescence by the energy that this reaction releases.

Molecular hydrogen can react with many elements and compounds, but at room temperature the reaction rates are usually so low as to be negligible. This apparent inertness is in part related to the very high dissociation energy of the molecule. At elevated temperatures, however, the reaction rates are high.

Explosive mixtures

Sparks or certain radiations can cause a mixture of hydrogen and chlorine to react explosively to yield hydrogen chloride, as represented by the equation $H_2 + Cl_2 \rightarrow 2HCl$. Mixtures of hydrogen and oxygen react at a measurable rate only above 300° C, according to the equation $2H_2 + O_2 \rightarrow 2H_2O$. Such mixtures containing 4 to 94 percent hydrogen ignite when heated to 550°–600° C or when brought into contact with a catalyst, spark, or flame. The explosion of a 2:1 mixture of hydrogen and oxygen is especially violent. Almost all metals and nonmetals react with hydrogen at high temperatures. At elevated temperatures and pressures hydrogen reduces the oxides of most metals and many metallic salts to the metals. For example, hydrogen gas and ferrous oxide react, yielding metallic iron and water, $H_2 + FeO \rightarrow Fe + H_2O$; hydrogen gas reduces palladium chloride to form palladium metal and hydrogen chloride, $H_2 + PdCl_2 \rightarrow Pd + 2HCl$.

Hydrogen is absorbed at high temperatures by many transition metals (scandium, 21, through copper, 29; yttrium, 39, through silver, 47; hafnium, 72, through gold, 79); and metals of the actinide (actinium, 89, through lawrencium, 103) and lanthanide series (lanthanum, 57, through lutetium, 71) to form hard, alloy-like hydrides. These are often called interstitial hydrides because, in many cases, the metallic crystal lattice merely expands to accommodate the dissolved hydrogen without any other change.

Hydrogen bond. Some covalently bonded hydrides have a hydrogen atom bound simultaneously to two separate electronegative atoms, which are then said to be hydrogen bonded. The strongest hydrogen bonds exist between the small, highly electronegative atoms of fluorine (F), oxygen, and nitrogen. In the bifluoride ion, HF_2^-, the hydrogen atom links two fluorine atoms. In the crystal structure of ice, each oxygen atom is surrounded by four other oxygen atoms, with hydrogen atoms between them. Some of the hydrogen bonds are broken when ice melts, and the structure collapses with an increase in density. Hydrogen bonding is important in biology because of its major role in determining the configurations of molecules. The helical (spiral) configurations of certain enormous molecular chains, as in proteins, are held together by hydrogen bonds. Extensive hydrogen bonding in the liquid state explains why hydrogen fluoride (HF), water (H_2O), and ammonia (NH_3) have boiling points much higher than those of their heavier analogues, hydrogen chloride (HCl), hydrogen sulfide (H_2S), and phosphine (PH_3). At the higher boiling temperatures, thermal energy is available to break up the hydrogen bonds and to permit vaporization.

Oxidation potential

The hydrogen in a strong acid, such as hydrochloric (HCl) or nitric (HNO_3), behaves quite differently. When these acids dissolve in water, hydrogen in the form of a proton, H^+, separates completely from the negatively charged ion, the anion (Cl^- or NO_3^-), and interacts with the water molecules. The proton is strongly attached to one water molecule (hydrated) to form the oxonium ion (H_3O^+), which in turn is hydrogen-bonded to other water molecules, forming species with formulas such as $H(H_2O)_n^+$ (the subscript n indicates the number of H_2O molecules involved). The oxidation of hydrogen (oxidation is the chemical change in which an atom loses one or more electrons) can be represented as the half reaction: $\frac{1}{2}H_2 \rightarrow H^+ + e^-$. The energy needed to bring about this reaction can be expressed as an oxidation potential. The oxidation potential for hydrogen is taken by convention to be zero, and all metals with positive oxidation potentials—*i.e.,* metals that are more easily oxidized (*e.g.,* zinc; $Zn \rightarrow Zn^{2+} + 2e^-$, 0.763 volt)—can, in principle, displace hydrogen from a strong acid solution: $Zn + 2H^+ \rightarrow Zn^{2+} + H_2$. Metals with negative oxidation potentials (*e.g.,* silver; $Ag \rightarrow Ag^+ + e^-$, −0.7995 volt) are inert toward the aqueous hydrogen ion.

ISOTOPES OF HYDROGEN

Discovery of deuterium, ^2H, or D

By means of the mass spectrograph he had invented, Francis William Aston in 1927 observed that the line for hydrogen corresponded to an atomic weight on the chemical scale of 1.00756. This value differed by more than the probable experimental error from the value based on the combining weights of hydrogen compounds, 1.00777. Other workers showed that the discrepancy could be removed by postulating the existence of a hydrogen isotope of mass 2 in the proportion of one atom of ^2H to 4,500 atoms of ^1H. The problem interested the U.S. chemist Harold C. Urey, who from theoretical principles predicted a difference in the vapour pressures of hydrogen (H_2) and hydrogen deuteride (HD) and thus the possibility of separating these substances by distillation of liquid hydrogen. In 1931 Urey and two collaborators detected deuterium by its atomic spectrum in the residue of a distillation of liquid hydrogen. Deuterium was first prepared in pure form by the electrolytic method of concentration: when a water solution of an electrolyte, such as sodium hydroxide, is electrolyzed, the hydrogen formed at the cathode contains a smaller fraction of deuterium than the water, and thus deuterium is concentrated in the residue. Almost pure deuterium oxide (D_2O, heavy water) is obtained when the solution is reduced to 0.00001 of its original volume. Deuterium can be concentrated also by the fractional distillation of water and by various chemical exchange reactions such as the following (g and l indicate gaseous and liquid states, respectively): $H_2O(g) + HD(g) \rightleftharpoons HDO(g) + H_2(g)$; $HDO(g) + H_2S(g) \rightleftharpoons HDS(g) + H_2O(g)$; $NH_3(l) + HD(g) \rightleftharpoons NH_2D(l) + H_2(g)$.

Creation of tritium, ^3H, or T

Tritium was first prepared in 1935 by bombarding deuterium (in the form of deuterophosphoric acid) with high-energy deuterons (deuterium nuclei):

$$^2D + {}^2D \longrightarrow {}^1H + {}^3T.$$

Tritium is present in minute concentrations in natural water. It is formed continuously in the upper atmosphere by cosmic-ray-induced nuclear reactions. Cosmic rays, consisting mainly of high-energy protons, react with nitrogen atoms to form neutrons, which in turn react with more nitrogen atoms to form tritium:

$$^{14}N + {}^1H \longrightarrow {}^{14}O + {}^1n \qquad {}^{14}N + {}^1n \longrightarrow {}^{12}C + {}^3T.$$

This naturally formed tritium ends up in the form of water and reaches the surface of the Earth in rain. Tritium is radioactive; it has a half-life of 12.5 years, decaying to a very soft (low energy) negative beta particle (electron; the positive beta particle is called a positron) and a helium-3 nucleus. When a sample of water is stored, it gradually loses its tritium because of radioactive decay. Thus by analyzing water for its tritium content, it is possible to elucidate details of water circulation among oceans, the atmosphere, rivers, and lakes. Tritium is made artificially in nuclear reactors by the reaction of thermal neutrons with lithium:

$$^6Li + {}^1n \longrightarrow {}^3T + {}^4He.$$

Corresponding compounds of the hydrogen isotopes differ slightly in their physical properties. This difference is shown by the properties of the waters, listed in Table 11, and of the elements, listed in Table 10. The same is true of their chemical properties, both thermodynamic and kinetic. Both deuterium and tritium are useful as isotopic tracers for the investigation of chemical structures and of reaction mechanisms. Generally the value of a tracer arises from the fact that, although its difference in mass or its radioactivity permits its detection, it is essentially active in the same way that the ordinary atoms of the element are. For most elements, a change of one or of a few mass units is such a small percentage of the total mass that the chemical differences between isotopes are negligible. For hydrogen, however, chemical reactions involving the different isotopes proceed at measurably different rates. These kinetic-isotope effects can be utilized in detailed studies of reaction mechanisms. The rates of reactions of compounds containing deuterium or tritium are usually less than those of the corresponding compounds of ordinary hydrogen.

Table 11: Physical Properties of the Waters

	H_2O	D_2O	T_2O
Density at 25° C in g/ml	0.99707	1.10451	—
Melting point, °C	0	3.81	4.49
Boiling point, °C	100	101.41	—
Temperature of maximum density, °C	3.98	11.21	13.4
Maximum density in g/ml	1.00000	1.10589	1.21502

The replacement of hydrogen by deuterium in biological systems can markedly alter the delicately balanced processes. It has been established that neither plants nor animals continue to live and thrive in water containing deuterium oxide in high concentrations.

Energy from deuterium and tritium

Deuterium and tritium are of interest in connection with thermonuclear (fusion) reactions. The explosion of a hydrogen bomb involves the collision and fusion of light nuclei, including deuterium and tritium. Should a method be found for controlling such fusion processes, as was done with the fission process of the earlier atomic bomb, the raw material for a practically unlimited supply of energy would be available in the deuterium content of water.

Deuterium oxide is useful in nuclear reactors as a moderator to slow down but not appreciably capture neutrons. It has the advantage of being a liquid that absorbs neutrons only slightly.

PRODUCTION AND APPLICATIONS OF HYDROGEN

The most important industrial method for the production of hydrogen is the catalytic steam–hydrocarbon process, in which gaseous or vaporized hydrocarbons are treated with steam at high pressure over a nickel catalyst at 650°–950° C to produce carbon oxides and hydrogen: $C_nH_{2n+2} + nH_2O \rightarrow nCO + (2n + 1)H_2$; $C_nH_{2n+2} + 2nH_2O \rightarrow nCO_2 + (3n + 1)H_2$. The primary reaction products are processed further in various ways, depending on the desired application of the hydrogen.

Another important process for hydrogen production is the noncatalytic partial oxidation of hydrocarbons under elevated pressures: $C_nH_{2n+2} + (n/2)O_2 \rightarrow nCO + (n + 1)H_2$. This process requires a feed system for delivering precise rates of fuel and oxygen, burners of special design to give rapid mixing of the reactants, a refractory-lined reactor, and a cooling system to recover heat from the effluent gases. The latter process is exothermic (heat producing), in contrast to the endothermic (heat absorbing) steam–hydrocarbon process.

In a third process, called the pressure catalytic partial oxidation method, the two preceding processes are combined to maintain the required reaction temperature without external heating of the catalyst bed. Superheated steam and hydrocarbons are mixed, preheated, and blended with heated oxygen in a diffuser at the top of the catalytic reactor. The oxygen reacts with the hydrocarbons in a space above the catalyst. The reactants then pass through a bed of nickel catalyst in which the steam–hydrocarbon reactions proceed almost to equilibrium.

Before 1940 most of the world production of hydrogen was made by processes based on coal or coke, the principal one being a water–gas reaction between steam and red-hot coke: $H_2O + C \rightarrow CO + H_2$. By 1970, however, relatively little hydrogen was being produced by such processes. For many years relatively small amounts of hydrogen had been produced by the electrolysis of aqueous solutions of salt or sodium hydroxide, the electrode reaction being $H_2O + e^- \rightarrow \frac{1}{2}H_2 + OH^-$. The reaction between sulfuric or hydrochloric acid and an active metal like zinc is utilized to liberate hydrogen in the laboratory, but such hydrogen usually contains trace quantities of volatile hydrides, such as arsine (AsH_3) and phosphine (PH_3), produced by impurities in the metal. These volatile impurities may be removed by bubbling the mixture of gases through a solution of a strong oxidizing agent, such as potassium permanganate.

A commercial method was developed for separating hydrogen from carbon monoxide synthesis gas by diffusion. The gas flows under pressure through bundles of tiny hollow polyester fibres through whose walls the hydrogen passes.

Use in ammonia manufacture

The largest single use of hydrogen in the world is in ammonia manufacture, which consumes about two-thirds of the world's hydrogen production. Ammonia is manufactured by the so-called Haber process, in which hydrogen and nitrogen react in the presence of a catalyst at pressures around 1,000 atmospheres and temperatures around 500° C: $N_2 + 3H_2 \rightarrow 2NH_3$. Large amounts of hydrogen are used in the preparation of methanol by the reaction $CO + 2H_2 \rightarrow CH_3OH$. This process is carried out in the presence of certain mixed catalysts containing zinc oxide and chromium oxide at temperatures between 300° and 375° C and at pressures between 275 and 350 atmospheres.

Another major application of hydrogen is in the catalytic hydrogenation of organic compounds. Unsaturated vegetable and animal oils and fats are hydrogenated to make margarine and vegetable shortening. Hydrogen is used to reduce aldehydes, fatty acids, and esters to the corresponding alcohols. Aromatic compounds can be reduced to the corresponding saturated compounds, as in the conversion of benzene to cyclohexane and of phenol to cyclohexanol. Nitro compounds can be reduced easily to amines.

Use as rocket fuel

Hydrogen has been used as a primary rocket fuel for combustion with oxygen or fluorine and is favoured as a propellant for nuclear-powered rockets and space vehicles. Another increasing use of hydrogen is in the direct reduction of iron ores to metallic iron and in the reduction of the oxides of tungsten and molybdenum to the metals. A hydrogen (reducing) atmosphere is employed in the pouring of special castings, in the manufacture of magnesium, in the annealing of metals, and for the cooling of large electric motors. Hydrogen was once used for inflating lighter-than-air vessels, such as dirigibles and balloons, but now helium is generally used because it is nonflammable. The barrage balloons used in England during World War II, however, were filled with hydrogen. Liquid hydrogen is used in the laboratory to produce low temperatures.

ANALYSIS

When atoms are excited, as in an electric discharge, they radiate light at discrete wavelengths that appear as lines in the spectrum. Inasmuch as the wavelengths of atomic spectral lines are characteristic of the element, the atomic spectrum may be used for identifying the element. The simplest of all such spectra is that of hydrogen. Johann Jakob Balmer, a Swiss mathematician and secondary school teacher, in 1885 discovered an equation for representing the wavelengths of hydrogen spectral lines, of which nine had been observed in the laboratory and of which five more were photographed in the spectrum of the star Sirius. The wavelengths, lambda (λ), in angstroms, were represented by the formula: $\lambda = 3645.6\ [m^2/(m^2 - 4)]$, m taking the successive values 3, 4, 5, etc. It was not until 1913 that a theoretical basis for this empirical relation was given by the Danish physicist Niels Bohr in his theory of atomic radiation.

The spinning motion of the proton gives it magnetic properties and causes it to precess in an applied magnetic field, much as a spinning top precesses in a gravitational field. The frequency at which a particular proton precesses is determined by its local electrical environment and by the strength of the applied magnetic field. When hydrogen compounds are irradiated with electromagnetic waves of a particular frequency, the phenomenon of resonance absorption occurs at magnetic field strengths that are different for each structurally (magnetically) distinguishable proton in the compound. Thus, proton magnetic resonance makes it possible to distinguish the structural types of hydrogen atoms present; furthermore, the absorption peak intensities are proportional to the number of hydrogen atoms of each kind. The absorption peaks are often split, however, because of the magnetic interaction of the proton magnetic moments among themselves. Proton-magnetic-resonance measurements provide data for the investigation of chemical structure.

One method for determining the total hydrogen content of a substance is to oxidize the substance completely in a stream of pure oxygen, which reacts with the hydrogen to produce water vapour. The resulting vapours are passed through a powerful dehydrating agent, such as magnesium perchlorate, which absorbs the water. From the increase in weight of the absorption tube containing the desiccant, the amount of hydrogen oxidized can be calculated. Gaseous hydrogen or hydrogen compounds may be oxidized by passing them over hot copper oxide, and the resulting water can then be collected and weighed and the amount of hydrogen calculated; to measure the hydrogen gas itself, the water vapour from the oxidation may be reduced to hydrogen gas by passing it over hot uranium metal—the hydrogen then being measured in a simple device called a gas buret.

Strongly acidic hydrogen atoms (as in compounds such as HCl, HNO_3, H_2SO_4, etc.) can be determined in solution by adding measured amounts of a strong base, such as sodium hydroxide, NaOH, until the acid is neutralized, using an indicator to determine the end point. The net reaction is $H^+ + OH^- \rightarrow H_2O$. Weakly acidic hydrogen atoms (such as that attached to the oxygen in ethanol, C_2H_5OH, and those attached to the nitrogen in acetamide, CH_3CONH_2) can be converted to methane (measured in a gas buret) by reaction with the methyl Grignard reagent, CH_3MgI. Hydridic hydrogen atoms (as in $NaBH_4$, LiH, etc.) can be converted to molecular hydrogen (measured in a gas buret) by reaction with an aqueous acid.

(W.L.J./Ed.)

Alkali metals

The alkali metals are the elements in Group Ia of the periodic table of the elements (see Figure 10). In order of increasing atomic number they are lithium (Li), sodium (Na), potassium (K), rubidium (Rb), cesium (Cs), and francium (Fr). Sodium and potassium are the sixth and seventh most abundant of the elements, comprising respectively 2.6 and 2.4 percent of the Earth's crust. The other alkali metals are considerably more rare, with rubidium, lithium, and cesium respectively forming 0.03, 0.007, and 0.0007 percent of the Earth's crust. Francium, a natural radioactive isotope, is very rare and was not discovered until 1939. Because of their high reactivity, the alkali metals are never found as free metals in their natural state. They generally are found combined with other elements in the form of simple or complex compounds. The simpler compounds of the alkali metals are soluble in water and therefore are easily extracted and subjected to chemical operations for purposes of separation and purification. Minerals belonging to this class, such as halite (sodium chloride, NaCl), sylvite (potassium chloride, KCl), and carnallite (a potassium-magnesium chloride, $KCl \cdot MgCl_2 \cdot 6H_2O$), although somewhat rare, are the most important commercial sources of the alkali metals; the more complex, water-soluble minerals are far more abundant in the Earth's crust.

The alkali metals have all of the physical properties generally associated with metals, including silver-like lustre, high ductility, and excellent conductivity of electricity and heat. Lithium is the lightest metallic element. The alkali metals are low melting, ranging from a high of 179° C for lithium to a low of 28.5° C for cesium. Alloys of alkali metals exist that melt as low as −78° C.

High reactivity

The alkali metals are extremely reactive and combine readily with most of the substances found in the atmosphere. (Only lithium, however, reacts with nitrogen.) The alkali metals all react vigorously, and often violently, with water, releasing hydrogen and forming strong caustic solutions. Most common nonmetallic substances such as the halogens, halogen acids, sulfur, and phosphorus react with the alkali metals. The alkali metals themselves react with many organic compounds, particularly those containing an active hydrogen atom (that is, one that can be replaced readily).

Sodium is by far the most important alkali metal in terms of industrial use. It is employed in the reduction of organic compounds, in the preparation of sodium peroxide and sodium cyanide, and in the manufacture of tetraethyl lead. As a free metal, it is used as a heat transfer fluid in nuclear reactors. Its compounds, of the widest industrial importance and manufactured in hundreds of thousands of tons annually, include common salt (NaCl), baking soda ($NaHCO_3$), soda ash (Na_2CO_3), and caustic soda (NaOH). Potassium has considerably less use than sodium as a free metal. Potassium salts, however, are consumed in considerable tonnages in the manufacture of fertilizers. Lithium metal is used in certain light-metal alloys, and it is employed as a reactant in organic syntheses. Lithium also has potential use as a battery anode in electrically propelled automobiles. Rubidium and cesium and their compounds have but limited use.

HISTORY

Saltpetre

Alkali metal salts were known to the ancients. The Old Testament refers to a salt called *neter* (sodium carbonate), which was extracted from the ash of vegetable matter. Saltpetre (potassium nitrate) was used in gunpowder, which was introduced into Europe in about the 12th century. Saltpetre was also one of the chief ingredients of the incendiary mixture called Greek fire.

In October 1807 the English chemist Sir Humphry Davy isolated potassium and then sodium, the discoveries coming only a few days apart. The term sodium is derived from the Italian *soda*, a term applied in the Middle Ages to all alkalies; potassium comes from the French *potasse*, a name used for the residue left in the evaporation of aqueous solutions derived from wood ashes.

Lithium was discovered by the Swedish chemist J.A. Arfvedson in 1817 while analyzing the mineral petalite. The name lithium is derived from *lithos*, the Greek word for stony. The element was not isolated in pure form until Davy produced a minute quantity by the electrolysis of lithium chloride.

While the German chemists Robert Bunsen and Gustav Kirchhoff were investigating the mineral waters in the Palatinate in 1860, they obtained a filtrate that was characterized by two lines in the blue region of its spectrum (the light emitted when the sample was inserted into a flame). They suggested the presence of a new alkali element and called it cesium, derived from the Latin *caesius*, used to designate the blue of the sky. The same researchers, on extracting the alkalies from the mineral lepidolite, separated another solution, which yielded two spectral lines of violet colour. They proposed the name rubidium for the element in this solution from the Latin *rubidus*, which was used for the darkest red colour. Francium was not discovered until 1939 by Marguerite Perey of the Curie Institute in Paris.

Electrolytic production

In the 19th century the only use for the alkali metals was the employment of sodium as a reagent in the manufacture of aluminum. When the electrolytic process for aluminum purification was established, it appeared that large-scale use of sodium would cease. Subsequent improvements in the electrolytic production of sodium, however, reduced the cost of this element to such an extent that it could be employed economically in processes such as the preparation of sodium peroxide. The chief improvement in the electrolytic process was the introduction of continuous electrolysis of sodium hydroxide, a technique called the

Castner process. The Castner process was used until the evolution of the Downs-cell process in 1926. This process, in which a molten sodium chloride–calcium chloride mixture is electrolyzed, has largely supplanted the older process in the production of sodium.

GENERAL PROPERTIES OF THE GROUP

Physical properties. Metals are ordinarily characterized by their high thermal and electrical conductivity and by their lustre, ductility, and malleability. The alkali metals as a class exhibit these properties to a high degree. Chemically, each of the alkali metals has a valence of one (*i.e.,* one atom of an alkali metal combines with one atom of another univalent element, such as fluorine); this property is a consequence of their having only a single electron in the outermost shell of their atoms. In passing from lithium to francium, this single electron tends to be less strongly held. Generally, the energy necessary to remove the outermost electron from the atoms of an element, the ionization energy, decreases in the periodic table toward the left and downward in each vertical file, with the result that the most easily ionizable element in the entire table is francium, followed closely by cesium. The alkali metals, which make up the extreme left-hand file, have ionization energies ranging from 124.3 kilocalories per mole (Li) to 89.7 kcal/mole (Cs) (omitting the rare radioactive element francium). The alkaline-earth metals, the next group to the right, have higher ionization energies, ranging from 214.9 (beryllium) to 120.1 kcal/mole (barium).

Electronegativities

The electronegativity scale of the elements compares the ability of the atoms of the various elements to attract electrons to themselves. In the periodic table the electronegativities range from 0.7 for cesium, the least electronegative of the elements, to 4.0 for fluorine, the most electronegative. Metals are ordinarily considered to be those elements having values less than 2.0 on the electronegativity scale. As a group the alkali metals are the least electronegative of the elements, ranging from 0.7 to 1.0 on the scale, while the alkaline earths, the next group on the table, have electronegativities ranging from about 0.9 to 1.5.

The relatively simple electronic structure of the atoms of the alkali metals (a single electron in the outermost shell of each) results in a considerable degree of regularity and predictability in their physical and chemical properties. Bonds of the alkali metals with all nonmetals are essentially ionic (resulting from mutually attractive forces existing between positive and negative electrical charges). The physical properties vary in a regular pattern and, to a considerable extent, can be attributed to the effects of the mobile outermost electron.

Table 12 summarizes the important physical and thermodynamic properties of the alkali metals. Generally, the mobile electrons in metals are responsible for their high thermal and electrical conductivity and their high heat capacity; this is especially true of the alkali metals. These metals are all characterized by a body-centred cubic crystallographic arrangement (a standard pattern of atoms in their crystals), with eight nearest neighbours to each atom.

The closest distance between atoms, a characteristic property of crystals, increases with increasing atomic weight of the alkali metal atoms. As a group, the alkali metals have a looser crystallographic arrangement than any of the other metallic crystals, and cesium, because of its greater atomic weight, has an interatomic distance that is greater than that of any other metal.

Vapour-pressure data for the alkali metals and for two alloys formed between elements of the group are given in Figure 11, which shows that the vapour pressures increase in regular fashion with increasing atomic weight.

Melting points and boiling points

The melting points of the alkali metals as a group are lower than those of any other nongaseous group of the periodic table, ranging between 179° C for lithium and 28.5° C for cesium. Only mercury among the metallic elements has a lower melting point than cesium. The low melting points of the alkali metals are a direct result of the large interatomic distances in their crystals and the weak bond energies associated with such loose arrays. These same factors are responsible for the low densities, low heats of fusion, and small changes in volume upon fusion of the metals. Lithium, sodium, and potassium are less dense than water.

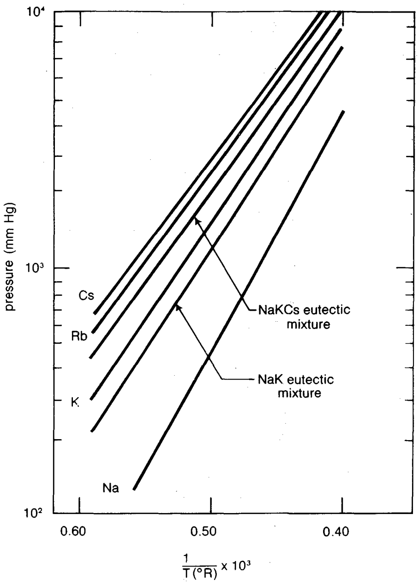


Figure 11: Vapour pressure in millimetres of mercury (mm Hg) plotted against the inverse of absolute temperature on the Rankine scale (T [°R]), for the alkali metals sodium (Na), potassium (K), rubidium (Rb), and cesium (Cs). As the atomic weight increases toward cesium the vapour pressure is higher at any temperature chosen. Alloys of metals that have single melting points, called eutectic mixtures, follow the same rule.

From J. Mansteller, F. Tepper, and S. Rodgers, *Alkali Metal Handling and Systems Operating Techniques* (1967); Gordon & Breach

Cesium is the most volatile of the alkali metals, with a boiling point of 671° C. The boiling points of the alkali metals increase in regular fashion as the atomic numbers decrease, the highest, 1,317° C, being that of lithium.

Chemical properties. Since the alkali metals are the most electropositive (the least electronegative) of elements, they react with a great variety of nonmetals.

Reactions with oxygen. All of the alkali metals react with oxygen—lithium and sodium to form monoxides and the heavier alkali metals to form superoxides. The rate of reaction with oxygen, or with air, depends upon whether the metals are in the solid or liquid state, as well as upon the degree of mixing of the metals with the oxygen, or air. In the liquid state alkali metals can be ignited in air with ease, generating copious quantities of heat and a dense choking smoke of the oxide.

In its chemical reactivity lithium more closely resembles Group IIa of the periodic table than it does the other metals of its own group. It is less reactive than the other alkali metals with water, oxygen, and halogens and more reactive with nitrogen, carbon, and hydrogen.

The free energy of formation (a measure of stability) of the alkali metal oxides at 25° C varies widely from a high of −133 kilocalories per mole for lithium oxide to −63 kilocalories per mole for cesium oxide. The close approach of the small lithium ion to the oxygen atom results in the unusually high free energy of formation of the oxide. Normal oxide formed on combustion of lithium is the simple oxide (Li_2O); it also is possible to make the peroxide (Li_2O_2) by passing oxygen through a liquid-ammonia solution of lithium metal. Sodium peroxide (Na_2O_2) can be similarly made, although the commercial method is by oxidation of sodium monoxide (Na_2O) with oxygen. Sodium superoxide (NaO_2) can be prepared with high oxygen pressures, whereas the superoxides of rubidium, potassium, and cesium can be prepared directly by combustion in air. By contrast, no superoxides have been isolated in pure form in the case of the alkaline-earth metals, although the heavier members of that group can be oxidized to the peroxide state. The ozonides of potassium, rubidium, and cesium, which are less stable than the lower

Table 12: Some Properties of the Alkali Metals

	lithium	sodium	potassium	rubidium	cesium	francium
Atomic number	3	11	19	37	55	87
Atomic weight (or stablest isotope)	6.94	22.99	39.1	85.47	132.91	(223)
Colour of element	silver	silver	silver	silver	silver	—
Melting point (°C)	179	97.8	63.65	38.89	28.5	—
Boiling point (°C)	1317	892	753.9	688	671	—
Density at 20° C (g/cm^3)	0.534	0.97	0.862	1.53	1.89	—
Volume increase on melting (percent)	1.51	2.63	2.81	2.54	2.66	—
Valence	1	1	1	1	1	1
Electronic configuration	$1s^2 2s^1$	$1s^2 2s^2 2p^6 3s^1$	(Ar)$4s^1$	(Kr)$5s^1$	(Xe)$6s^1$	(Rn)$7s^1$
Isotopic abundance (terrestrial, percent)	^{7}Li (92.6), ^{6}Li (7.4)	^{23}Na (100)	^{39}K (93.1), ^{41}K (6.88), ^{40}K (0.0119)	^{85}Rb (72.15), ^{87}Rb (27.85)	^{133}Cs (100)	—
Colour imparted to flame	red	yellow	violet	dark red	blue	—
Main spectral emission lines (wavelength, Å)	6708, 6104	5890, 5896	7699, 7665	4216, 4202	4593, 4555	—
Heat of fusion (cal/mole)	690	622	598	540	520	—
Specific heat (liquid, cal/g °C)	1.05	0.33	0.188	0.0880	0.0572	—
Electrical resistivity at 0° C (microhm-cm)	8.55	4.2	6.15	2.5	20.0	—
Magnetic susceptibility (cgs unit)	14.2×10^{-6}	16.0×10^{-6}	20.8×10^{-6}	17.0×10^{-6}	29.0×10^{-6}	—
Crystal structure	BCC*	BCC*	BCC*	BCC*	BCC*	—
Radius						
Atomic (Å)	1.52	1.85	2.31	2.44	2.62	—
Ionic (M$^+$, Å)	0.68	0.97	1.33	1.47	1.67	1.80
Metallic (single bond, Å)	1.225	1.572	2.025	2.16	2.35	—
Ionization energy (first, kcal/mole)	124.3	118.4	100	96.3	89.7	—
Oxidation potential at 25° C ($M \rightarrow M^+ + e^-$, volts)	3.04	2.71	2.92	2.92	2.92	—
Electronegativity	1.0	0.9	0.8	0.8	0.7	0.7

*Body-centred cubic.

oxides, can be prepared by the reaction of the superoxides with ozone.

Reactions with water. The alkali metals all react violently with water. The rate of the reaction depends on the degree of metal surface presented to the liquid. With small metal droplets or thin films of alkali metal, the reaction can be explosive. The rate of the reaction of water with the alkali metals increases with increasing atomic weight of the metal. With the heavier alkali metals the hydroxides are highly soluble; thus they are removed readily from the reacting surface, and the reaction can proceed with unabated vigour. The reaction involves equimolar mixtures (that is, equal numbers of atoms or molecules) of the alkali metal and water to form a mole (an amount equal to that of the reactants) of alkali metal hydroxide and half a mole of hydrogen gas. These reactions are highly exothermic (give off heat), and the hydrogen that is generated can react with oxygen to further increase the heat that is generated.

Reactions with nonmetals. Of the alkali metals, only lithium reacts with nitrogen, and it forms a nitride (Li_3N). In this respect it is more similar to the alkaline-earth metals than to the Group Ia metals. Lithium also forms a relatively stable hydride, whereas the other alkali metals form hydrides that can be more readily dissociated. Lithium forms a carbide (Li_2C_2) similar to that of calcium. The other alkali metals do not form stable carbides, although they do react with the graphite form of carbon to give intercalation compounds (substances in which the metal atoms are inserted between layers of carbon atoms in the graphite structure).

The alkali metals can be burned in atmospheres of the various halogens to form the corresponding halides. The reactions are highly exothermic, producing up to 235 kcal/mole for lithium fluoride. The alkali metals react with nonmetals in Groups Va and VIa of the periodic table. Sulfides can be formed by the direct reaction of the alkali metals with elemental sulfur, furnishing a variety of sulfides. Phosphorus combines with the alkali metals to form phosphides with the general formula M_3P.

Formation of alloys. The characteristics of alloy behaviour in alkali metals can be evaluated in terms of the similarity of the elements participating in the alloy. Elements with similar atomic volumes form solid solutions (that is, mix completely in all proportions); some dissimilarity in atomic volumes results in eutectic-type systems (solutions formed over limited concentration ranges); and further dissimilarity results in totally immiscible systems.

The elements potassium, rubidium, and cesium, which have rather similar atomic volumes and ionization energies, form complete solid solutions and mixed crystals. Sodium, which is a significantly smaller atom than potassium and has a higher ionization energy, tends to form eutectic systems with potassium, rubidium, and cesium. Even greater dissimilarity exists in the atomic volumes of sodium and lithium, resulting in insolubilities of the liquid phases. The consolute temperature (the temperature at which the two liquids become completely miscible) increases on going from the lithium–sodium alloy system to the lithium–cesium system. Lithium and cesium can coexist as two separate liquid phases at temperatures up to at least 1,100° C.

Eutectics

There is only one example of solid miscibility in alkali–alkaline-earth-metal binaries—the lithium–magnesium system, in which the two elements are very similar. Sodium forms compounds only with barium in the alkaline-earth-metal series. The heavier alkali metals all tend to form immiscible liquid phases with the alkaline earths.

Several of the elements of the Group IIb of the periodic table (zinc, cadmium, and mercury) react with the alkali metals to form compounds. Mercury forms at least six compounds, commonly termed amalgams, with each of the five alkali metals; and, with the exception of the amalgam with lithium, the highest melting point compound in each series has the formula MHg_2. Lithium and sodium also form compounds with cadmium and zinc.

INDIVIDUAL ALKALI METALS

Lithium. *Occurrence and production.* Lithium occurs as salts in mineral springs, and its concentration in seawater is 0.1 parts per million. Crude phosphate, containing 20 percent lithia, or lithium oxide (Li_2O), is recovered from brines at Searles Lake, California. Lithium is found in pegmatite ores, such as spodumene ($LiAlSi_2O_6$) and lepidolite (of varying structure), or in amblygonite ($LiAlFPO_4$) ores, with Li_2O contents ranging between 4 and 8.5 percent.

Method of manufacture

Lithium is manufactured by the electrolysis of a fused mixture of lithium and potassium chlorides. The lithium chloride consumed in the electrolysis may be obtained from lithium-containing silicate minerals by a number of different processes. Spodumene, for example, is calcined (heated), treated with sulfuric acid, and then neutralized with calcium carbonate to form lithium sulfate solution. Soda ash (sodium carbonate) is added to produce lithium carbonate, the basic material from which other lithium compounds are derived. Lithium chloride may be volatilized from aluminosilicate materials by heating with carbon in a stream of chlorine. Lithium chloride also may be prepared in high yield by passing hydrogen chloride over lepidolite ore at 935° C.

The addition of potassium chloride to lithium chloride allows electrolysis to occur at 400°–420° C. Since the voltage at which decomposition of lithium chloride takes place is lower than that of potassium chloride, lithium is deposited in purities greater than 97 percent. Graphite anodes are used in the electrolytic production of lithium, while the cathodes are made of steel. The pure lithium formed at the cathode coalesces at the surface of the electrolyte to form a molten pool, which is protected from reaction with air by a thin film of the electrolyte. The lithium is ladled from the cell and cast by pouring it into a mold at a temperature only slightly above the melting point, leaving the solidified electrolyte behind. The solidified lithium is then remelted, and materials insoluble in the melt either float to the surface or sink to the bottom of the melt pot. The remelting step reduces the potassium content to less than 100 parts per million (ppm).

Many lithium alloys are produced directly by the electrolysis of molten salts, containing lithium chloride in the presence of a second chloride, or by the use of cathode materials that interact with the deposited lithium, introducing other elements into the melt.

Significant uses. The isotope lithium-6, the less common of the two chief isotopes, yields tritium gas when bombarded with neutrons. Tritium so produced is employed in the manufacture of hydrogen bombs, among other uses. The principal industrial applications for lithium are in metallurgy, where the active element is used as a scavenger (remover of impurities) in the refining of such metals as iron, nickel, copper, and zinc, and their alloys. A large variety of nonmetallic elements are scavenged by lithium, including oxygen, hydrogen, nitrogen, carbon, sulfur, and the halogens. Lithium is utilized to a considerable extent in organic synthesis. Organolithium compounds act somewhat like organomagnesium compounds and undergo reactions similar to the Grignard reaction, a standard synthetic procedure in organic chemistry. Lithium compounds also serve as catalysts in synthetic rubber manufacture.

Lithium has potential value as a heat-transfer fluid for high power-density nuclear reactors. The lithium-7 isotope, the more common stable isotope, has a low nuclear cross section (that is, absorbs neutrons very poorly) and is an ideal primary coolant for nuclear reactors in which coolant temperatures above about 800° C are required.

Lightweight lithium–magnesium alloys have structural applications in the aerospace and other industries. Metallic lithium is used in the preparation of compounds such as lithium hydride.

Chemical properties. In many of its properties lithium exhibits the same characteristics as do the more common alkali metals sodium and potassium; thus, the metal is highly reactive and forms strong hydroxide solutions. In many respects lithium also shows similarities to the elements of the alkaline-earth group, especially magnesium, which has similar atomic and ionic radii. This similarity is seen in oxidation properties, the monoxide being normally formed in each case.

Nuclear properties. Lithium, which exhibits no natural radioactivity, has two isotopes of mass number 6 and 7 respectively. The ^{7}Li : ^{6}Li ratio is between 12 and 13. The isotopes ^{8}Li (half-life 0.855 sec) and ^{9}Li (half-life 0.17 sec) have been produced by nuclear bombardment. The ^{7}Li isotope has a low nuclear cross section (0.003 barn) permitting consideration of the substance as a coolant for nuclear reactors in such mobile systems as satellites and spacecraft.

Biological properties. The widespread occurrence of lithium in plants results in a wide, although low-level, distribution of lithium in animals. Lithium salts have complex effects when absorbed into the body. They are not highly toxic, although they can affect body processes dependent upon salt balance. Lithium poisoning can occur, particularly with patients on low sodium diets. Lithium salts that are ingested are absorbed and are rapidly excreted in the urine.

The most abundant of the alkali metals

Sodium. *Occurrence and production.* Sodium is the most abundant of the alkali metals. Sodium chloride (table salt) is the most common compound of sodium, but many others also are known. Sodium is an important constituent of a number of silicate materials, such as feldspars and micas. There are huge deposits of rock salt in various parts of the world, and sodium nitrate deposits exist in Chile and Peru. The sodium content of the sea is approximately 1.05 percent, corresponding to a concentration of approximately 3 percent of sodium halides. Sodium has been identified in both the atomic and ionic forms in the spectra of stars and the interstellar medium. Analysis of meteorites indicates that the silicate material present has an average content of approximately 4.6 atoms of sodium for every hundred atoms of silicon.

Sodium is the alkali metal of most commercial importance by far. Most processes for the production of sodium involve the electrolysis of sodium chloride. The element is used chiefly to produce tetraethyl lead, which increases the antiknock rating of gasoline.

The principal sodium compounds of commerce are the chloride, carbonate, and sulfate. Large quantities of sodium chloride are, for example, employed in the production of other heavy (industrial) chemicals.

Significant uses. Two of the earliest uses of sodium were in the manufacture of sodium cyanide and sodium peroxide, both uses that continue in importance. A present major use for sodium is in the manufacture of tetraethyl lead by means of the reaction of lead–sodium alloy with ethyl chloride. Substantial amounts of sodium are also used in the manufacture of sodium alkyl sulfates for the synthetic-detergent industry, in which the sodium alkyl sulfates are employed as the principal ingredient.

Sodium also is used as a starting material in the manufacture of sodium hydride. In addition, sodium is employed in the production of dyes and dye intermediates, in the synthesis of perfumes, and in a wide variety of organic reductions. It is used in the purification of hydrocarbons and in the polymerization of unsaturated hydrocarbons. In many organic applications, sodium is used in the form of dispersions in hydrocarbon liquid media.

Heat-transfer uses of sodium

Sodium is an excellent heat-transfer fluid, and because of this property, it has potentially large-scale uses. Fast-breeder nuclear reactors, cooled by sodium, may generate a portion of the world's electrical power by the year 2000. Sodium forms an alloy with potassium (NaK) which melts at −10° C. At one time NaK was considered for use as the heat-transfer fluid in submarine reactor systems. Later, sodium was investigated as the secondary coolant system in a high power-density reactor system for space use. The alloy also is used for cooling crucibles in consumable arc-melting processes for preparing titanium.

Because of the excellent electrical conductivity of sodium, the substance has been considered as a substitute for copper in electrical conductors. It is used extensively in metallurgy as a deoxidant and as a reducing agent for the preparation of calcium, zirconium, titanium, and other transition metals. Commercial production of titanium involves reduction of titanium tetrachloride with sodium.

Chemical properties. Generally, elemental sodium is more reactive than lithium, and in aqueous solution it reacts to form a stronger base than LiOH. Its chemistry is well explored.

Reaction with oxygen

Sodium is ordinarily quite reactive with air, and the reactivity is a function of the relative humidity, or water-vapour content, of the air. The corrosion of solid sodium by oxygen also is accelerated by the presence of small amounts of impurities in the sodium. In ordinary air, sodium metal reacts to form a sodium hydroxide film, which can rapidly absorb carbon dioxide from the air, forming sodium bicarbonate. Sodium does not react with nitrogen. It is significantly more reactive in air as a liquid than as a solid, and the liquid can ignite at about 125°

C. In a comparatively dry atmosphere, sodium burns quietly, giving off a dense, white, caustic smoke, which can cause choking and coughing. The temperature of burning sodium increases rapidly to more than 800° C, and under these conditions the fire is extremely difficult to extinguish.

Sodium monoxide (Na_2O) is ordinarily formed upon oxidation of sodium in dry air. The superoxide (NaO_2) can be prepared by heating metallic sodium to 300° C in an autoclave (heated pressure vessel) containing oxygen at high pressure. Another route to the superoxide is oxidation of sodium peroxide treated to have a large surface area.

Sodium that is heavily contaminated with the monoxide may be readily purified by filtration since the solubility of the oxide in molten sodium is low. This low solubility is utilized to a considerable extent in continuous purification processes of the sodium in large liquid-metal reactor systems. A second technique for removing the oxide, called cold trapping, involves running the molten sodium through a cooled, packed bed of material, upon which the oxide can precipitate. Filtration and cold trapping also are effective in removal of gross quantities of carbonate, hydroxide, and hydride.

Reaction with water and hydrogen

The reaction with water of liquid sodium having a high surface area can be explosive. The sodium–water reaction is highly exothermic (that is, heat is given off):

$$Na + H_2O \longrightarrow NaOH + {}^1/_2\,H_2 + 33.67 \text{ kilocalories per mole at } 25^\circ \text{ C.}$$

Tests have indicated, however, that sodium and water cannot be mixed fast enough to produce the shock waves characteristic of high explosives. The explosive hazards of the reaction are associated primarily with the hydrogen gas that is formed.

Pure sodium begins to absorb hydrogen appreciably at about 100° C, the rate of absorption increasing with the temperature. Pure sodium hydride can be formed at temperatures above 350° C by exposing sodium to hydrogen gas at a high flow rate. The dissociation of sodium hydride is considerably greater than that of lithium hydride but slightly less than that of potassium hydride.

Reactions with nonmetals

Generally, alkali metals react with halogen gases, the degree of reactivity decreasing with increasing atomic weight of the halogen. Sodium is no exception to this statement. Under certain conditions of reaction, sodium and halogen vapours react to produce light (chemiluminescence). Halogen acids, such as hydrochloric acid, react vigorously with sodium, yielding the sodium halides. The reactions are highly exothermic, with heats of reaction (energy given off) of −71.8 and −76.2 kilocalories respectively for the reactions with hydrofluoric and hydrochloric acids. Sodium is attacked by other strong mineral acids to form the corresponding salts. It reacts with the fumes of nitric acid at 15° C to form sodium nitrate and with acetic and sulfuric acids to form sodium acetate and sodium sulfate. With molten sulfur it reacts violently to produce polysulfides; under more controlled conditions it reacts with organic solutions of sulfur. Liquid selenium and tellurium both react vigorously with solid sodium to form selenides and tellurides.

Sodium shows relatively little reactivity with carbon, although the existence of lamellar (layer-like) materials prepared from graphite with the formula NaC_{64} has been reported. At 625° C carbon monoxide reacts with sodium to form sodium carbide and sodium carbonate.

With the exception of the oxides of the Group IVb metals (titanium, zirconium, and hafnium), the oxides of the transition metals are all reduced to the respective metals with elemental sodium. Sodium also reacts with a large number of metallic halides, displacing the metal from the salt and forming a sodium halide in the process. This reaction is used in the preparation of several of the transition metals themselves, including titanium and tantalum.

Reactions in liquid ammonia

Sodium dissolves in liquid ammonia to give intense blue solutions, and at ordinary temperatures a slow reaction occurs in such solutions, resulting in the formation of sodamide, $NaNH_2$, and hydrogen. Conversion of alkali metals to their amides takes several days in the case of lithium and sodium, but for the heavier alkali metals conversion occurs in a half-hour or less—and with cesium it takes only 15 minutes. The reaction of sodium and liquid ammonia can be catalyzed by the addition of any of a number of metal and metal oxide catalysts.

Liquid ammonia often is used as a solvent for sodium, allowing a number of reactions to occur at ordinary temperatures that would otherwise need heat. Sodium superoxide (NaO_2), for example, can be formed by passing oxygen through ammonia solutions of sodium at −77° C. Ammonia also serves as a solvent for reactions of sodium with arsenic, tellurium, antimony, bismuth, and a number of other low-melting metals.

Organic reactions

The organic reactions of sodium have been studied to a greater extent than those of any of the other alkali metals. Sodium reacts with anhydrous alcohols to form the respective alcoholates. The reaction is most vigorous with methanol and decreases with increasing molecular weight of the alcohol. Sodium methoxide is produced on an industrial scale by reaction of sodium with excess methanol. Organic acids react with sodium to form sodium salts.

The high free energy of formation of sodium halides permits the dehalogenation of a large number of organic halides, the formation of the sodium halide being energetically favoured. The so-called Wurtz reaction—based on this principle—is used in organic synthesis to a considerable extent:

$$2RCl + 2Na \rightarrow R\text{–}R + 2NaCl.$$

By this reaction, octane can be made from bromobutane and sodium. Organosodium compounds include a number in which the sodium atom is bonded directly to a carbon atom; an example is methylsodium, $Na\text{–}CH_3$. Such compounds can be prepared by the action of sodium on mercury dialkyls or diaryls, as in the following equation:

$$Hg(CH_3)_2 + 2Na \rightarrow 2NaCH_3 + Hg.$$

Sodium reacts violently with a number of halogenated hydrocarbons. A violent explosion occurs when a mixture of carbon tetrachloride and sodium is subjected to shock, and even when the sodium is diluted to a considerable extent—as in sodium amalgam—a brisk reaction with carbon tetrachloride occurs.

Alloys

Sodium is completely miscible with the alkali metals below it in the periodic table (potassium, rubidium, and cesium). A eutectic (that is, an alloy that melts lower than its components) melting at −10° C is formed in the sodium–potassium system and is known commercially as NaK. Its composition is approximately 78 percent potassium, and it is used as a heat-transfer fluid and as an organic reactant. The eutectics formed in the sodium–rubidium and sodium–cesium binary systems melt respectively at −4.5° and −30° C. Sodium is the minor component with potassium and cesium of the ternary alloy NaKCs, melting at −78° C. This fluid is the lowest melting liquid alloy yet isolated.

Sodium also forms alloys with the alkaline-earth metals. Beryllium is soluble in sodium only to the extent of a few atomic percent at approximately 800° C. Liquid sodium and magnesium are only partially miscible. The degree of solubility in sodium of the alkaline-earth metals increases with increasing atomic weight, with the result that the solubility of calcium is 10 percent by weight at 700° C. In the sodium–strontium system, there is a considerable degree of miscibility. Sodium forms a number of compounds with barium, and several eutectics exist in the system.

The precious metals, silver, gold, platinum, palladium, and iridium, and the white metals, such as lead, tin, bismuth, and antimony, alloy to an appreciable extent with liquid sodium. Cadmium and mercury also react with sodium, and a number of compounds exist in both binary systems. Seven sodium–mercury compounds, or amalgams, exist, with Hg_2Na having the highest melting point (354° C). Sodium amalgams are used chiefly for carrying out reactions in situations in which pure elemental sodium would be violently reactive and difficult to control. The solubility of transition metals in alkali metals is generally very low, often in the one to ten parts per million range even at temperatures in excess of 500° C.

The sodium compounds of major commercial importance are the chloride, hydroxide, carbonate, sulfate, and, to a lesser extent, the peroxide, sulfide, and halides other than the chloride.

Nuclear properties. In its nuclear properties, sodium is a simple element, with but one natural isotope, ^{23}Na, exhibiting no natural radioactivity. Several isotopes have been prepared by nuclear reactions, with the ^{22}Na isotope having the longest half-life—2.6 years. The ^{24}Na isotope (half-life 15 hours) is produced by irradiation in a nuclear reactor. Because of this reaction a sodium-cooled reactor must have a second heat-transfer loop so that radioactive sodium does not come in contact with the environment.

Biological properties. Sodium salts, particularly sodium chloride, are found almost everywhere in biological material. Sodium is an essential element for life, as is potassium, and the two elements maintain a definite balance within the cell structure. Most of the biological effects of sodium salts are the result of the cation (sodium ion), with the anion (negative ion) apparently not playing a dominant role.

The presence of salinity in soils is often detrimental to plant growth. Sodium ions replace calcium and other ions in clay complexes, transforming the clay to a sticky mass; water percolation is then drastically reduced, and the basicity of the soil rises markedly.

Sodium as a dietary requirement

The tolerance of fish to changes in salinity is often quite remarkable. Many marine bacteria and diatoms are able to tolerate salt concentrations as great as 25 percent. The minimum sodium requirement for mammals appears to be 0.05 percent of the diet, corresponding in a normal man to a requirement of one to two grams of salt per day, resulting in an average sodium content of body tissues of 0.24 percent. There is a wide variation of sodium content in the different tissues, with whole blood containing approximately 0.62 percent sodium chloride, whereas skin has a sodium content of less than 0.1 percent. There is a relationship between salt content and water balance of the body, a low intake of salt causing loss of water. Considerable quantities of sodium are lost through the skin by sweating, and considerable quantities can be excreted in the urine.

Potassium. *Occurrence and production.* Igneous rocks, shale, and sediment are minerals containing much of the potassium that is found in nature. The potassium content of the Dead Sea is estimated at approximately 1.7 percent potassium chloride, and many other salty bodies of water are rich in potassium. The waste liquors from certain salt works may contain up to 40 grams per litre of potassium chloride and are used as a source of potassium. The salt deposits in Stassfurt, Germany, which are rich in potassium sulfate and chloride, are one of the most important sources of potassium in the world. Potassium also is found in nature as the nitrate.

Extraction and purification of potassium salts

The techniques used for extracting potassium salts depend to a considerable extent on the nature of the ore body and the character of the foreign salts mixed with the potassium salt. The Stassfurt salt deposit contains large deposits of carnallite, $MgCl_2 \cdot KCl \cdot 6H_2O$. These deposits became so important with respect to potash (potassium hydroxide) production that Germany had a virtual monopoly on production of that substance up to the time of World War I. The other chief sources of potash are in France, Austria, Spain, India, and Chile. Seawater, brines, many lake deposits, and the ashes of vegetation also are used as sources of potash.

Potassium chloride can be separated from carnallite ore by preferential precipitation of potassium chloride during evaporation of the mother liquor. In the Searles Basin, California, deposit, brine is evaporated in vacuum pans, and the sodium chloride and sulfate are deposited. The resulting liquor is boiled to a point where potassium chloride is separated, producing a crude potash deposit containing about 66 percent potassium chloride.

Production of potassium

Potassium is produced by sodium reduction of molten potassium chloride at 870° C. Molten KCl is continuously fed into a packed distillation column, while sodium vapour is passed up through the column. By condensation of the volatile potassium at the top of the distillation tower, the reaction

$$Na + KCl \rightarrow K + NaCl$$

is forced to the right. Efforts to devise a scheme for the electrolytic production of potassium have been unsuccessful because there are few salt additives that can reduce the melting point of potassium chloride to temperatures where electrolysis is efficient.

Significant uses. Most of the potassium produced is converted to the superoxide directly by combustion in dry air. The superoxide is used as a source of oxygen in respiratory equipment, since it generates oxygen and also absorbs carbon dioxide.

Sodium–potassium alloy (NaK) is used only to a limited extent as a heat-transfer coolant, although it has significant potential for this purpose. It is presently being evaluated for use in crucible cooling in titanium arc melting. The alloy also is used as an organic catalyst.

Chemical properties. The chemical properties of potassium are similar to those of sodium, although the former is considerably more reactive. Potassium differs from sodium in a number of respects. Whereas sodium is essentially unreactive with graphite, potassium, rubidium, and cesium react to form a series of interlamellar compounds, the richest in metal having the formula MC_8. Compounds are formed with carbon–alkali metal atomic ratios of 8, 16, 24, 36, 48, and 60 : 1. The graphite lattice is expanded during penetration of the alkali metal between the layers. Potassium reacts with carbon monoxide at temperatures as low as 60° C to form an explosive carbonyl ($K_6C_6O_6$), a derivative of hexahydroxybenzene. Sodium is essentially nonreactive with carbon monoxide.

Liquid potassium and NaK both are more reactive than liquid sodium with air and oxygen. Potassium superoxide (KO_2) is the preferred oxide, whereas sodium forms the monoxide (Na_2O). Oxidation of NaK produces Na_2O preferentially. The superoxide of potassium is a yellow solid consisting of $K^+$and O_2^-ions. It also can be formed by oxidation of potassium amalgam with dry air or oxygen.

Potassium reacts violently with water to produce one-half mole of hydrogen per mole of potassium and water and generating approximately 47 kcal/gram-atom of heat. It can be stored under nitrogen gas with no reaction. It reacts with hydrogen at approximately 350° C to form the hydride.

Reactions with halogens

Potassium is highly reactive with halogens and detonates when it contacts liquid bromine. Violent explosions also have been observed when mixtures of potassium and halogen acids are subject to shock. Explosions also have occurred when potassium is being mixed with a number of metal halide salts or with organic-halogen compounds.

At elevated temperatures, potassium reduces carbon dioxide to carbon monoxide and carbon. Solid carbon dioxide and potassium react explosively when subjected to shock. Oxidation of potassium amalgam with carbon dioxide results in the formation of potassium oxalate. Potassium is not reactive with benzene, although heavier alkali metals such as cesium add across the double bonds between carbon atoms in the benzene molecules, giving organometallic products.

In addition to the alloys of potassium with lithium and sodium (see above), alloys with other alkali metals are known. Complete miscibility exists in the potassium–rubidium and potassium–cesium binary systems. The latter system forms an alloy melting at approximately −38° C. Modification of the system by the addition of sodium results in a ternary eutectic melting at approximately −78° C. The composition of this alloy is 3 percent sodium, 24 percent potassium, and 73 percent cesium. Potassium is essentially immiscible with all of the alkaline-earth metals, as well as with zinc, aluminum, and cadmium.

Nuclear properties. The two stable isotopes of potassium are ^{39}K and ^{41}K, which together comprise 99.99 percent of the element as it occurs in nature, about 13.5 times as much of the former isotope normally being present. The natural radioactivity of potassium is due to beta radiation from the ^{40}K isotope (10^9 years half-life), which is present in the element in nature to the extent of about 0.01 percent. The disintegration of ^{40}K is used in geological age calculations.

Biological properties. Potassium is an essential ingredient in all forms of life. Each organism has a closely maintained potassium level and a relatively fixed potassium–sodium ratio. Potassium is the primary inorganic cation within the living cell, and sodium is the most abundant cation in extracellular fluids. Potassium is an essential ingredient of fertilizers. The potassium content of plants varies considerably, but it is ordinarily in the range of 0.5 to 2 percent. In man, the ratio of potassium between the cell and plasma is approximately 27 : 1. The potassium content of muscle tissue is approximately 0.3 percent, whereas that of blood serum is about 0.01 to 0.02 percent. The dietary requirement for normal growth is approximately 3.3 grams of potassium per day, but the ingestion of more than 20 grams of potassium results in distinct physiological effects. Excess potassium is excreted in the urine, and a significant quantity may be lost during sweating.

Rubidium and cesium. Because rubidium and cesium are very similar in their physical and chemical properties, they are conveniently discussed together.

Occurrence and production. Rubidium and cesium often occur together in nature. Rubidium, however, is more widely scattered and seldom forms a natural mineral; it is found only as an impurity in other minerals, ranging in content up to 5 percent in such minerals as lepidolite and pollucite. Brine samples have been analyzed that contain up to six parts per million of rubidium.

The Earth's crust is estimated to contain only 7 ppm of cesium, whereas it contains 310 ppm of rubidium. Pollucite ($Cs_4Al_4Si_9O_{26} \cdot H_2O$), however, is a cesium-rich mineral resembling quartz. It contains 40.1 percent cesium on a pure basis, and impure samples are ordinarily separated by hand-sorting methods to greater than 25 percent cesium. Large pollucite deposits have been found in Rhodesia and in the lithium-bearing pegmatites at Bernic Lake, Manitoba. Cesium also occurs in a rare mineral called rhodizite, and it is found in low concentrations in lepidolite and in salt brines and saline deposits.

Source of rubidium

The principal commercial source of rubidium is accumulated stocks of a mixed carbonate produced as a by-product in the extraction of lithium salts from lepidolite. Primarily a potassium carbonate, the by-product also contains approximately 23 percent rubidium and 3 percent cesium carbonates. The primary difficulty associated with the production of either pure rubidium or pure cesium is that these two elements are always found together in nature and also are mixed with other alkali metals; because these elements are very similar chemically, their separation presents numerous problems. Before development of procedures based on thermochemical reduction and fractional distillation, the elements were purified in the salt form through laborious crystallization techniques. Once pure salts have been prepared by precipitation methods, it is a relatively simple task to convert them to the free metal. This end is ordinarily accomplished by calcium reduction in a high-temperature vacuum system in which the highly volatile alkali metal is distilled from the reaction mixture.

Direct reduction of the mixed carbonates from lepidolite purification, followed by fractional distillation, is perhaps the most important of the commercial methods for producing rubidium. The mixed carbonate is treated with excess sodium at approximately 650° C, and much of the rubidium and cesium passes into the metal phase. The resulting crude alloy is vacuum distilled to form a second alloy considerably richer in rubidium and cesium. This product is then refined by fractional distillation in a tower to produce elemental rubidium more than 99.5 percent pure.

The direct thermochemical reduction of pollucite ore with sodium metal is the primary commercial source of cesium metal. Use of this process may lead to significant decreases in cesium prices if the market for cesium increases. In the process ground pollucite ore is treated with elemental sodium at approximately 650° C to form a sodium–cesium alloy containing some rubidium impurity. Fractional distillation of this alloy in a distillation tower at approximately 700° C produces 99.9 percent pure cesium metal.

Significant uses. Rubidium metal has few commercial uses, and it is of very minor economic significance. The high prices and uncertain and limited supply of the metal discourage the development of any uses. Only a few commercial uses have been developed for the metal itself, although the salts are of slightly greater economic significance because of their use in photocells and radio vacuum tubes. Cesium metal also is produced in rather limited amounts because of its relatively high cost. Cesium has been evaluated for power systems for space applications and as a potential rocket fuel for interplanetary travel. Because the cesium atom can be ionized thermally and the positively charged ions accelerated to great speeds, cesium systems can provide specific impulses for rocket propulsion that are extraordinarily high. Cesium also has application in thermionic converters that generate electricity directly within nuclear reactors. Another potentially large application of cesium metal is in the production of low-melting NaKCs eutectic alloy. Cesium salts are used in moderate quantities in the manufacture of photomultiplier tubes, infrared lamps, scintillation counters, and spectrophotometers.

Space applications of cesium

Chemical properties. Rubidium and cesium are difficult to handle because they react spontaneously in air. If the surface area of a sample of the metal is large, it can burn to form superoxides. Rubidium superoxide (RbO_2) is a yellow powder; cesium superoxide has a more reddish cast. Peroxides of both metals (Cs_2O_2 and Rb_2O_2) can be formed by oxidation of the metal with the required amount of oxygen. Rubidium forms two other oxides (Rb_2O and Rb_2O_3), but the reactions of cesium with oxygen are much more complex.

Rubidium and cesium are miscible in all proportions and have complete solid solubility; a melting-point minimum of +9° C is reached. Both rubidium and cesium form a number of mercury amalgams. Because of the increased specific volumes of rubidium and cesium, as compared to the lighter alkali metals, there is less tendency for them to form alloy systems with other metals.

Nuclear properties. Two rubidium isotopes are found in nature. The stable rubidium-85 makes up 72 percent of the mixture. Radioactive rubidium-87 makes up 28 percent; it emits beta rays with a half-life of about 6×10^{11} years. Cesium has only one naturally occurring isotope, with a mass number of 133. A number of cesium isotopes with short half-lives have been produced artificially by nuclear reactions.

Francium. The existence of francium was predicted by Mendeleyev in his periodic classification of the elements. In 1939 the French chemist Marguerite Perey discovered the radioactive element, which has beta activity and a half-life of 22 minutes (^{223}Fr). Nineteen other isotopes of francium also have been reported. It has been calculated that the total quantity of francium in the Earth's crust is only 24.5 grams. Information on biological and chemical aspects of the element is virtually nonexistent.

ANALYTICAL CHEMISTRY OF THE ALKALI METALS

Chemical separation of lithium can be effected by the use of the organic nitrogen-containing compound picrolonic acid, which precipitates the heavier alkali metals from chloride solutions, leaving lithium behind. Lithium can be detected with a number of azo dyes that give highly sensitive colour reactions with lithium ions in alkaline solution. A modification of the uranyl acetate test (the precipitation of an insoluble sodium salt with uranyl acetate) is used as a standard test for the presence of sodium. The use of a cobaltinitrite solution permits separation of potassium from sodium by precipitation of the insoluble potassium salt. There are essentially no satisfactory analytical methods for rubidium and cesium based on the use of reagents in solution; these elements are commonly determined spectroscopically—that is, by measurements of their flame spectra or light-emission patterns.

Instrumental analysis

Ordinary methods of chemical analysis generally require extensive preliminary separation. With the alkali metals this process frequently can be avoided by the use of flame tests, which are quickly and easily applied and able to detect small amounts of the alkali metals in the presence of other elements. When salt solutions of the alkali metals are atomized in the coal gas burner of a simple flame photometer, which measures emitted light and must be provided with a suitable interference filter (to remove unwanted emissions), the metals can be determined generally in the parts-per-million range. Determination of one alkali metal in the presence of another, however, can result in interference, which can be reduced using specially prepared standard solutions containing known amounts of the interfering metals for comparative purposes.

Emission spectrographic analysis (measuring light emitted at particular wavelengths, a more precise method than by a photometer) is a useful method for analyzing trace quantities of one alkali metal in another and is also very helpful for analyzing metallic impurities. Colorimetric methods, which compare colours produced in solution, are often used for the analysis of metallic impurities.

The analysis of the alkali-metal samples for the presence of nonmetallic elements, such as oxygen, carbon, hydrogen, and nitrogen, requires specialized techniques. The oxygen content of sodium and potassium samples can be determined by extraction of the free alkali metal with mercury, leaving behind mercury-insoluble oxides and carbonates, which can subsequently be analyzed using solution methods. The oxygen content of rubidium and cesium can be accurately determined by precise measurement of the freezing point of these two elements.

The carbon content of alkali metals can be analyzed by oxidation of the alkali metal in pure oxygen, followed by infrared measurement of the carbon dioxide generated during combustion. For the analysis of nitride in lithium, the nitride commonly is converted to ammonia, and the ammonia is measured by colorimetric analysis.

The sampling of alkali metals to obtain representative examples for the analysis of nonmetallic elements is a difficult art. Such sampling ordinarily is carried out by melting the whole batch of alkali metal and extracting a sample by dipping a bucket into the melt, withdrawing it, and analyzing the contents of the bucket. (F.T./Ed.)

Alkaline-earth metals

The term alkaline-earth metals is the generic name for the chemical elements beryllium (Be), magnesium (Mg), calcium (Ca), strontium (Sr), barium (Ba), and radium (Ra), which are the members of group IIa of the periodic table (see Figure 10) of the elements. Prior to the 19th century, substances that were nonmetallic, insoluble in water, and unchanged by fire were known as earths. Those earths, like lime, that resembled the alkalies (soda ash and potash) were designated alkaline earths. Alkaline earths were thus distinguished from the alkalies and from other earths, such as alumina and the rare earths. By the first decade of the 19th century it became clear that the earths, formerly considered to be elements, were in fact oxides, compounds of a metal and oxygen. The metals whose oxides make up the alkaline earths then came to be known as the alkaline-earth metals and have been classified in group II of the periodic table ever since Mendeleyev proposed his first table in 1869.

Chemical and physical properties

The alkaline-earth metals are extremely electropositive; that is, like the alkali metals of group Ia, their atoms easily lose electrons to become positive ions (cations). Most of their typical compounds are therefore ionic: salts in which the metal occurs as uniformly divalent cations, M^{2+}, where M represents any group IIa atom. The salts are colourless unless they include a coloured anion (negative ion). Typical alkaline-earth compounds, calcium chloride ($CaCl_2$) and calcium oxide (CaO), may be contrasted with the compounds of the uniformly monovalent alkali metals, sodium chloride (NaCl) and sodium monoxide (Na_2O). The oxides of the alkaline-earth metals are basic (*i.e.,* alkaline, in contrast to acidic). A fairly steady increase in electropositive character is observed in passing from beryllium, the lightest member of the group, to radium, the heaviest; as a result of this trend, beryllium oxide is only weakly basic and even shows acidic properties, whereas barium and radium oxide are strongly basic. The metals themselves are highly reactive reducing agents; that is, they readily give up electrons to other substances that are, in the process, reduced.

All the metals and their compounds find commercial application to some degree, especially magnesium alloys and a variety of calcium compounds. Magnesium and calcium, particularly the latter, are abundant in nature and play significant roles in geological and biological processes. Radium is a rare element; all its isotopes are radioactive.

HISTORY

The earliest known alkaline earth was lime (Latin *calx*), which is now known to be calcium oxide; it was used in ancient times in the composition of mortar. Magnesia, (the name derives probably from the ancient district of Magnesia in Asia Minor), the oxide of magnesium, was shown to be an alkaline earth different from lime by the Scottish chemist Joseph Black in 1755; he observed that magnesia gave rise to a soluble sulfate, whereas that derived from lime was known to be insoluble. In 1774 Carl Wilhelm Scheele, the Swedish chemist who discovered oxygen, found that the mineral called heavy spar or *barys* (Greek, heavy) contained a new earth, which became known as baryta (barium oxide). A further earth, strontia (strontium oxide), was identified by Adair Crawford, a London physician and chemist, in 1790 on examining a mineral (strontium carbonate) found in a lead mine at Strontian in Argyllshire, Scotland. Beryllia (beryllium oxide) was extracted from the mineral beryl and recognized as an earth by the French analytical chemist Louis-Nicolas Vauquelin in 1797. Though at first confused with alumina (aluminum oxide) because both dissolve in alkali, beryllia was shown to be distinct; unlike alumina, it reprecipitated when the alkaline solution was boiled for some time. Beryllia was originally called *glucina* (Greek *glykys,* sweet) because of its sweet taste. (This etymological root is retained in France, where the element beryllium is also known as glucinium.)

Magnesium, calcium, strontium, and barium, elements derived from alkaline earths, were isolated as impure metals by Sir Humphry Davy in 1808 by means of the electrolytic method he had previously used for the isolation of the alkali metals potassium and sodium. The alkaline-earth metals were later produced by reduction of their salts with free alkali metals, and it was in this way (the action of potassium on beryllium chloride) that beryllium was first isolated by the German chemist Friedrich Wöhler in 1828. Radium was discovered in 1898 by means of its radioactivity by Pierre and Marie Curie, who separated it with great difficulty from barium.

Discovery of alkaline-earth metals

PHYSICAL AND CHEMICAL BEHAVIOUR

The alkaline-earth elements are highly metallic and are good conductors of electricity. They have a gray-white lustre when freshly cut but tarnish readily in air, particularly the heavier members of the group. Beryllium is sufficiently hard to scratch glass, but barium is only slightly harder than lead. The melting points and boiling points of the group are higher than those of the corresponding alkali metals; they vary in an irregular fashion, magnesium having the lowest (mp 650° C and bp 1,105° C) and beryllium the highest (mp 1,283° and bp about 2,500°). The elements crystallize in one or more of the three regular close-packed metallic crystal froms. Chemically, they are all strong reducing agents. The free metals are soluble in liquid ammonia—the dark-blue solutions of calcium, strontium, and barium arousing considerable interest because they are thought to contain metal ions and the most unusual species, solvated electrons, or electrons resulting from the interaction of the metal and the solvent. Highly concentrated solutions of these elements have a metallic, copper-like appearance, and further evaporation yields residues containing ammonia, which correspond to the general formula $M(NH_3)_6$. The solutions are strong reducing agents and are useful in a number of chemical processes.

The atoms of the alkaline-earth elements all have similar electronic structures, consisting of a pair of electrons (designated *s* electrons) in an outermost orbital, within which is a stable electronic configuration corresponding to that of a noble gas. The noble gas elements—helium (He), neon (Ne), argon (Ar), krypton (Kr), xenon (Xe), and radon

Table 13: Some Properties of the Alkaline-Earth Metals

	beryllium	magnesium	calcium	strontium	barium	radium
Atomic number	4	12	20	38	56	88
Atomic weight (stablest isotope)	9.0122	24.312	40.08	87.62	137.34	(226)
Colour of element	gray	silvery white	silvery white	silvery white	silvery white	silvery white
Melting point (°C)	1,283	650	842–48	769	725	700
Boiling point (°C)	~2,500	1,105	1,487	1,384	1,640	<1,737
Density at 20° C (g/cm^3)	1.85	1.74 (5° C)	1.54	2.54	3.51	(about 5)
Oxidation number	2	2	2	2	2	2
Electronic configuration	$1s^2 2s^2$	(Ne)$3s^2$	(Ar)$4s^2$	(Kr)$5s^2$	(Xe)$6s^2$	(Rn)$7s^2$
Isotopic abundance (terrestrial, percent)	^{9}Be(100)	^{24}Mg(78.70), ^{25}Mg(10.13), ^{26}Mg(11.17)	^{40}Ca(96.97), ^{42}Ca(0.64), ^{43}Ca(0.145), ^{44}Ca(2.06), ^{46}Ca(0.0033), ^{48}Ca(0.18)	^{84}Sr(0.56), ^{86}Sr(9.86), ^{87}Sr(7.02), ^{88}Sr(82.56)	^{130}Ba(0.101), ^{132}Ba(0.097), ^{134}Ba(2.42), ^{135}Ba(6.59), ^{136}Ba(7.81), ^{137}Ba(11.32), ^{138}Ba(71.66)	—
Radioactive isotopes (mass numbers)	7, 10–11	20–23, 27–28	37–39, 41, 45, 47, 49, 50	80–83, 85, 89–93, 95	123, 125–29, 131, 133, 139–43	213–17, 219–30
Electrical resistivity at 0° C (microhm-cm)	4.0	4.45	3.91	23	60	—
Crystal structure*	HCP	HCP	FCC, HCP, BCC	FCC, HCP, BCC	BCC	—
Radius, ionic (M^{2+}, Å)	0.31	0.65	0.99	1.13	1.35	1.40
Diameter, atomic (Å) (coordination number of 12)	2.25	3.20	3.93	4.30	4.48	—
Ionization energy (electron volts)						
First	9.32	7.64	6.11	5.69	5.21	5.28
Second	18.21	15.03	11.87	11.03	10.00	10.14
Third	153.85	80.12	51.21	—	—	—
Electrode potential (according to international convention), $M^{2+} + 2e^- \rightarrow M$, at 25° C, volts	−1.85	−2.37	−2.87	−2.89	−2.90	−2.92
Electronegativity (Pauling)	−1.5	−1.2	−1.0	−1.0	−0.9	−0.9

*HCP = hexagonal close packed, FCC = face-centred cubic (cubic close packed), BCC = body-centred cubic.

Electronic configuration

(Rn)—have generally complete electron shells. Strontium has the configuration $1s^2 2s^2 2p^6 3s^2\ 3p^6 3d^{10} 4s^2 4p^6 5s^2$, which may be written as (krypton core) $5s^2$, or simply (Kr) $5s^2$. Similarly, Be may be designated as (He) $2s^2$, Mg as (Ne) $3s^2$, Ca as (Ar) $4s^2$, Ba as (Xe) $6s^2$, and Ra as (Rn) $7s^2$. The prominent lines in the atomic spectra of the elements, obtained when the elements are heated under certain conditions, arise from states of the atom in which one of the two *s* electrons has been promoted to a higher energy orbital. The *s* electrons are relatively easily ionized (removed from the atom), and this ionization is the characteristic feature of alkaline-earth chemistry. The ionization potential (the energy required to strip an electron from the atom) falls continuously in the series from beryllium (9.32 electron volts [eV]) to barium (5.21 eV); radium, the heaviest in the group, has a slightly higher ionization potential (5.28 eV). The small irregularities observed in the otherwise smooth change as one proceeds down the group as it appears in the periodic table are explained by the uneven filling of electron shells in the successive rows of the table. The *s* electrons may also be promoted to *p* orbitals of the same principal quantum number (within the same shell) by energies similar to those required to form chemical bonds; the atoms are, therefore, able to form stable covalently bonded structures, unlike helium, which has the otherwise analogous electronic configuration of $1s^2$.

Zinc, cadmium, and mercury, the group IIb elements, are often compared with the alkaline-earth elements calcium, strontium, and barium. Cadmium, for example, has the electronic configuration (Kr) $4d^{10}5s^2$, with the ten $4d$ electrons taking virtually no part in chemical bonding. The $5s^2$ electrons, however, are much less readily ionized in cadmium than they are in strontium, for the $4d$ electrons act as an ineffective shield for the corresponding increased charge on the cadmium nucleus. The chemistry of the IIb metals, therefore, is markedly less ionic than the chemistry of the alkaline-earth metals.

Ionic character. The chemistry of the alkaline-earth metals, like that of the alkali metals, is for the most part reasonably interpreted in terms of an ionic model for the compounds formed. This model is less satisfactory for the chemistry of beryllium and magnesium than for the heavier alkaline-earth metals. In fact, most beryllium compounds are molecular (covalent) rather than ionic. Although there is some evidence for the transient existence of singly charged alkaline-earth ions, in most cases the chemistry of these elements is dominated by the formation and properties of the doubly charged M^{2+} ions, in which the outermost *s* electrons have been stripped from the metal atom. The resulting ion is stabilized by electrostatic interaction with a solvent, like water, which has a high dielectric constant and a great ability to absorb electrical charge, or by combination with ions of opposite charge in an ionic lattice such as is found in salts. The extra energy required to remove the second *s* electron (the second ionization potential being approximately twice the first) is more than compensated for by the extra binding energy present in the doubly charged ion. The removal of a third electron from an alkaline-earth atom, however, would require an expenditure of energy greater than could be recouped from any chemical environment. As a result, therefore, the alkaline-earth metals show a constant valence of +2 in their compounds.

Ionic sizes

The sizes of the ions of the alkaline-earth elements increase steadily from Be^{2+}, which has a radius of 0.31Å or 31×10^{-10} cm, to Ra^{2+} with a radius of 1.40Å. The ionic radius of the europium ion Eu^{2+} (1.12Å) is very close to that of strontium Sr^{2+} (1.13Å); this means that Eu^{2+} ions can sometimes be used as a "probe" for the alkaline-earth metals, substituting for strontium ions in situations where advantage can be taken of the spectroscopic and magnetic properties that make Eu^{2+} readily identifiable. The ionic radius of the cadmium ion Cd^{2+} (0.97Å) is very similar to that of calcium Ca^{2+} (0.99Å). A quantitative comparison of cadmium and calcium chemistry, therefore, clearly shows up the less ionic character of cadmium chemistry without complications due to differences in ionic size. A related comparison may be made between mercury and strontium, because of the similar ionic radii of Hg^{2+} (1.10Å) and Sr^{2+} (1.13Å).

The chemistry of radium is less well investigated than that of the other alkaline-earth metals. As expected, however, it is in general an extrapolation of the chemistry of calcium, strontium, and, especially, barium.

The group IIa ions are readily hydrated, with the strength of bonding to the water molecules increasing with decreasing ionic radius. Although the number of water molecules directly attached to the metal ion may be greater with the larger ions for purely steric (geometrical) reasons, the total number of water molecules associated with the metal ion nevertheless increases inversely with the size of the ion itself (as shown by migration experiments conducted in aqueous solution). Large anions, such as sulfate, tend to form weak ion-pair complexes more readily with the larger metal ions of the family, but weak-acid anions, such as acetate, tend to form stronger complexes with the smaller metal ions, particularly those of magnesium and beryllium. That many of these complexes are molecular rather than ionic is shown by their ready extraction from aqueous solution (which preferentially dissolves ionic substances) into organic solvents (which dissolve molecular ones).

INDIVIDUAL ALKALINE-EARTH METALS

Beryllium. Beryllium is widely distributed in the Earth's crust, of which it composes about 0.001 percent; its cosmic abundance is 20 on the scale in which silicon, the standard, is one million. Although there are about 30 recognized minerals containing beryllium, only three—beryl ($3\,BeO \cdot Al_2O_3 \cdot 6SiO_2$), phenacite ($2\,BeO \cdot SiO_2$), and chrysoberyl ($BeO \cdot Al_2O_3$)—are of any significance, and only the first is of industrial importance. The precious forms of beryl, emerald and aquamarine, have a composition closely approaching that given above, but industrial ores contain less beryllium; most beryl is obtained as a by-product of other mining operations, the larger crystals being picked out by hand.

Exploration for beryllium has been considerably facilitated by the development of the berylometer, a portable device that detects very small quantities of beryllium in almost any mixture. The device depends on the bombardment of the sample with gamma rays from the radioactive isotope antimony-24. The gamma rays interact specifically with beryllium atoms to give neutrons that can be counted electronically.

The extraction of beryllium is complicated by the fact that beryllium oxide is found only as a minor constituent, tightly bound to alumina and silica. Treatment with acids, roasting with complex fluorides, and liquid–liquid extraction have all been employed to concentrate beryllium oxide. The oxide is converted to fluoride via ammonium beryllium fluoride and then heated with magnesium to form elemental beryllium. The element is purified by vacuum melting.

Uses of beryllium

Metallic beryllium is grayish, and its chemical properties somewhat resemble those of aluminum. Though resistant to air oxidation under normal conditions, it is readily attacked by acids and alkalies. The applications of beryllium depend upon its unusual combination of physical properties. A very light metal, it has a high mechanical strength and a high melting point. Because of its low atomic weight it has a high transparency to X-rays (17 times greater than aluminum), which makes it ideal as a material for windows of X-ray tubes. Its ability to slow down fast neutrons has found considerable application in atomic piles. Beryllium has been developed as a structural material for specific applications particularly within the aerospace industry.

Beryllium is used in a number of commmercial neutron sources. The alpha particles released by radioactive decay of radium atoms react with atoms of beryllium to give, among the products, neutrons with a wide range of energies—up to about 5×10^6 eV. If radium is encapsulated, however, so that none of the alpha particles reach beryllium, neutrons of energy less than 600,000 eV are produced by the more penetrating gamma radiation from the decay products of radium.

Magnesium. Magnesium, comprising about 2.5 percent of the Earth's crust, is the eighth most abundant element and, after aluminum and iron, the third most plentiful structural metal; its cosmic abundance is estimated as 9.1×10^5 atoms (Si = 10^6 atoms). It occurs as carbonates (magnesite, $MgCO_3$, and dolomite, $CaCO_3 \cdot MgCO_3$) and in many common silicates, including asbestos, talc, and olivine. It also is found as the hydroxide (brucite), chloride (carnallite, $KCl \cdot MgCl_2 \cdot 6H_2O$), and sulfate (kieserite).

Magnesium is commercially produced chiefly by the electrolysis of molten magnesium chloride and by the thermal reduction of magnesium oxide with ferrosilicon (see INDUSTRIES, EXTRACTION AND PROCESSING: *Magnesium production*).

It is the lightest metal that can be commonly used for structural purposes (its density is 65 percent that of aluminum and 22 percent that of iron), a property widely exploited. Its thermal and electrical conductivity and its melting point are very similar to those of aluminum. Whereas aluminum is attacked by alkalies but is resistant to most acids, magnesium is resistant to most alkalies but is readily attacked by most acids (chromic and hydrofluoric acids are important exceptions). At normal temperatures it is stable in air and water due to the formation of a thin protective skin of oxide (but burns rapidly when heated in air), and it is attacked by steam. Magnesium is a powerful reducing agent and is used to produce other metals from their compounds (*e.g.,* titanium, zirconium, and hafnium). It reacts directly with many elements; with organic halides in ether solution it forms Grignard reagents. Because of its ready combustibility, magnesium finds application in explosive and pyrotechnic devices.

Uses of magnesium

Magnesium is mainly used in the form of alloys, which usually contain 10 percent or less of other elements, generally added to increase the strength of the metal. The most important alloys are those of aluminum and zinc. Casting, rolling, extruding, and forging techniques are all employed with the alloys, and further fabrication of the resulting sheet, plate, or extrusion is carried out by normal forming, joining, and machining operations. Magnesium is the easiest structural metal to machine and has often been used when a large number of machining operations are required.

Magnesium is essential to all living systems. The photosynthetic function of plants depends upon the action of chlorophyll pigments, which contain magnesium at the centre of a complex, nitrogen-containing ring system (porphyrin). These magnesium compounds enable the energy of light to be used to convert carbon dioxide and water to carbohydrates and oxygen and thus directly or indirectly provide the key to nearly all living processes. Magnesium also takes part in a number of enzyme reactions, which control energy transfer in living cells.

Natural sources of calcium

Calcium. Calcium is the fifth most abundant element in the Earth's crust, to which it contributes an estimated 3.64 percent; its cosmic abundance is estimated at 4.9×10^4 atoms (Si = 10^6 atoms). One of the most widely distributed elements, it occurs as carbonate (chalk, limestone, marble, calcite), sulfate (anhydrite, gypsum), fluoride (fluorite or fluorspar), and phosphate (apatite). It is also found in a large number of silicates and aluminosilicates, in salt deposits, and in natural waters, including the sea. Calcium carbonate deposits dissolve in water that contains carbon dioxide to form calcium bicarbonate, $Ca(HCO_3)_2$. This process often results in the formation of caves, and may reverse to deposit limestone as stalactites and stalagmites. Calcium is essential to plant and animal life. Many living organisms concentrate calcium in their shells or skeletons, and indeed in higher animals calcium is the most abundant inorganic element. Many important carbonate and phosphate deposits owe their origin to living organisms.

The metal is produced by thermal reduction of lime with aluminum under high vacuum and by electrolysis of fused calcium chloride (see INDUSTRIES, EXTRACTION AND PROCESSING: *Production of other important industrial metals.*) It reacts with water and, on heating, with oxygen, nitrogen, hydrogen, halogens, boron, sulfur, carbon, and phosphorus; its commercial applications depend largely on these reactions. Although it compares favourably with sodium as a reducing agent, calcium is more expensive and less reactive. In many deoxidizing, reducing, degasifying, and alloying applications, however, calcium often is preferred because of its lower volatility. Small percentages of calcium are used in many alloys for special purposes.

The sulfate (as uncalcined gypsum) is employed as a soil corrector. Calcined gypsum is used in making tile, wallboard, lath, and various plasters. Plaster of Paris, the hemihydrate, $CaSO_4 \cdot \frac{1}{2} H_2O$, is produced by partial calcination at about 120° C; mixed with water, it forms a plastic mass that hydrates to a hard white plaster.

The dihydrogen sulfite, $Ca(HSO_3)_2$, is made by the action of sulfur dioxide on a slurry of $Ca(OH)_2$. Its aqueous solution under pressure dissolves the lignin in wood to leave cellulose fibres and thus finds considerable application in the paper industry.

The phosphates are the principal minerals for the production of phosphate fertilizers and for a whole range of phosphorus compounds. The rock is usually treated with sulfuric acid to form $Ca(H_2PO_4)_2$, which may be applied directly to the land. The fluoride, CaF_2, is important to the production of hydrofluoric acid, which is made from CaF_2 by the action of sulfuric acid.

Strontium. Strontium comprises about 0.025 percent of the Earth's crust; its cosmic abundance is estimated as 18.9 atoms ($Si = 10^6$ atoms). Although it is widely distributed with calcium, there are only two principal ores of strontium alone, celestite ($SrSO_4$) and strontianite ($SrCO_3$). Metallic strontium may be obtained by the electrolysis of strontium chloride or by reduction of the oxide with aluminum, but there is little commercial demand for the metal because calcium and barium are more readily obtainable and fulfill the functions for which strontium might be used.

The chemistry of strontium is quite similar to that of calcium. The biological properties of strontium are also very close to those of calcium and distinct from those of barium, whose soluble compounds, for example, are poisonous.

Hazards of strontium-90

The artificial isotopes strontium-89 (52-day half-life) and more especially strontium-90 (28-year half-life), found in radioactive fallout, are extremely hazardous. They become deposited in bones, where they replace calcium; their radiation then damages bone marrow, impairs the process of forming new blood cells, and may induce cancer.

Barium. Barium occurs to the extent of about 0.05 percent in the Earth's crust; its cosmic abundance is estimated as 3.7 atoms ($Si = 10^6$ atoms). Two minerals are common: the more important sulfate ($BaSO_4$, barite, barytes or heavy spar), used principally in oil-drilling "mud," and the carbonate ($BaCO_3$, witherite). The metal may be produced by electrolysis of barium chloride, but the most effective method is the reduction of the oxide by heating with aluminum or silicon in a high vacuum. Only a few tons of the metal are produced each year. It is used chiefly as a "getter" in vacuum tubes, where it removes the last traces of various gases.

Radium. About 16 isotopes of radium, all radioactive, are known; their half-lives, except for radium-226 (1,600 years) and radium-228 (5.8 years), are less than a few weeks. The long-lived radium-226 is found in nature as a result of its continuous formation from uranium-238 decay. Radium thus occurs in all uranium ores, but it is more widely distributed because it forms water-soluble compounds; the Earth's surface contains an estimated 1.8×10^{13} grams of radium.

The discovery of radium

The story of the discovery and isolation of radium by Pierre and Marie Curie (1898) is well known. The Curies observed that the major part of the radioactivity of pitchblende was caused by a previously unknown substance that could be concentrated with barium and then separated from it by laborious fractional crystallization. The separation was followed by the increase in intensity of new lines in the ultraviolet spectrum and by a steady increase in the apparent atomic weight of the material until a value of 225.18 was obtained, remarkably close to the accepted value of 226.03.

In modern technology, radium is separated from barium by fractional crystallization of the bromides, followed by purification by ion-exchange techniques for removal of the last 10 percent of the barium. Radium metal may be prepared by electrolytic reduction of its salts.

Radioactivity. The chemistry of radium is what would be expected of the heaviest of the alkaline earths, but the intense radioactivity is its most characteristic property. One gram of radium-226 undergoes 3.7×10^{10} disintegrations per second, producing energy equivalent to 6.8×10^{-3} calories, sufficient to raise the temperature of a well-insulated sample at the rate of 1° C every 10 seconds. The practical energy release is even greater than this, due to the production of a large number of short-lived radioactive decay products. The alpha particles emitted by radium may be used to initiate nuclear reactions; a mixture of beryllium (see above *Beryllium*) and radium is used as a neutron source. Radium mixed with a phosphor that could be excited by the alpha particles was widely used in the manufacture of luminous paints for watch and instrument dials until the early 1950s, but less hazardous alpha emitters have largely replaced it. Although radium-226 is not a very intense gamma emitter, its decay products are; thus radium has been used in medicine as a source of gamma rays for the irradiation of tumours. One of the products of radium decay is radon, the heaviest noble gas; this decay process is the chief source of that element.

Radium tends to concentrate in bone where alpha radiation interferes with red corpuscle production and can lead to bone cancer. The hazards of radium were disastrously emphasized by the deaths of a number of women who had been employed painting luminous dials during World War I and who had ingested considerable amounts of radium through the habit of licking the points of their brushes. The detection of exhaled radon provides a very sensitive test for radium absorption. (C.S.G.P./Ed.)

Zinc group elements

The zinc group elements consist of zinc (Zn), cadmium (Cd), and mercury (Hg), forming Group IIb of the periodic table of the elements (see Figure 10). They have properties in common, but they also differ in significant respects. All three are metals with a silvery-white appearance and relatively low melting points and boiling points; mercury is the only common metal that is liquid at room temperature, and its boiling point is lower than that of any other metal. The three are found in different proportions in the Earth's crust: it has been estimated that zinc is present to the extent of 80 parts per million (compared with 70 for copper and 16 for lead). The estimate for cadmium is only 0.15; commercially, it is always found associated with zinc or zinc–lead ores and is produced only as a by-product of zinc and lead smelting. The proportion of mercury in the Earth's crust is estimated at 0.08 parts per million. All important mercury deposits consist of mercuric sulfide, known as the mineral cinnabar.

Occurence in the Earth's crust

HISTORY

Zinc. Metallic zinc appeared much later in history than the other common metals. Copper, lead, tin, and iron can be obtained as the molten metals by heating their oxide ores with charcoal (carbon), a process called reduction, in shaft furnaces, which were developed quite early in history. Zinc oxide, however, cannot be reduced by carbon until temperatures are reached well above the relatively low boiling point of the metal (907° C). Thus, the furnaces developed to smelt the other metals could not produce zinc. Small quantities of metallic zinc can sometimes be found in the flues of lead blast furnaces. There is some evidence that the Greeks knew of the existence of zinc and called it *pseudargyras,* or "false silver," but they had no method of producing it in quantity. The Romans as early as 200 BC produced considerable quantities of brass, an alloy of zinc and copper, by heating in crucibles a mixture of zinc oxide and charcoal covered with lumps of metallic copper. The zinc oxide was reduced in the lower part of the crucible. Zinc vapour was formed and dissolved in the copper to form brass. At the end of the process the temperature was raised to melt the brass for casting into ingots. The realization that to make zinc it was necessary to produce the metal as a vapour and then condense it seems first to have been reached in India in the 14th century. In the West this principle was first applied in

Roman brass-making

England in 1743. At the end of the 18th century in Belgium and Poland improvements were made in the furnace, and the process remained unchanged until an electrolytic process was developed in 1917. At the end of the 1920s a radical advance was made in the United States by developing a continuous retort process, and during the 1930s an electrothermic process was designed for producing zinc continuously. A development of the 1960s was the zinc-lead blast furnace, in which rapid quenching of the gases is a key principle. Zinc production processes are treated in detail in INDUSTRIES, EXTRACTION AND PROCESSING: *Zinc production.*

Cadmium. Cadmium is a comparatively recent discovery, having been first found and described in 1817. Cadmium is present in small quantities in most zinc and lead ores, and it is produced entirely as an ancillary operation to zinc and lead smelting. The chief zinc ore, zinc blende, or sphalerite, consists mainly of zinc sulfide, containing from 0.1 to 0.3 percent cadmium. All methods of zinc production begin with the conversion of the sulfide into zinc oxide by roasting: the cadmium becomes concentrated in the fumes, which are treated in various steps until a product is obtained containing over 99.9 percent cadmium. Some lead ores also contain small quantities of cadmium, and if it is present in sufficient quantity it is recovered by a cycle of operations similar to that used by zinc smelters. Zinc producers who use the electrolytic process recover cadmium in a somewhat different way, but again the principle is the same, beginning with the roasting of zinc sulfide, followed by the treatment of the flue dusts.

Mercury. Mercury was known in Egypt and also probably in the East as early as 1500 BC. The name mercury originated in 6th-century alchemy, in which the symbol of the planet was used to represent the metal; the chemical symbol Hg derives from the Latin *hydrargyrum,* "liquid silver." Although its toxicity was recognized at an early date, its main application was for medical purposes.

Principal source of mercury

The chief commercial source of mercury is cinnabar, or mercury sulfide, which is mined in shaft or open-pit operations and refined by flotation. Over half the world supply of mercury comes from Spain and Italy. Most of the methods of extraction of mercury rely on the volatility of the metal and the fact that cinnabar is readily decomposed by air or by lime to yield the free metal. Because of the toxicity of mercury and the threat of rigid pollution control, attention is being directed toward safer methods of extracting mercury. These generally rely on the fact that cinnabar is readily soluble in solutions of sodium hypochlorite or sulfide, from which the mercury can be recovered by precipitation with zinc or aluminum or by electrolysis.

COMPARATIVE PROPERTIES OF THE GROUP

The electronic structure of the zinc group elements is shown in Table 14, and some of their properties are listed in Table 15.

Table 14: The Electron Structure of Zinc Group Elements

symbol	atomic number	shells					
		K	*L*	*M*	*N*	*O*	*P*
Zn	30	2	8	18	2		
Cd	48	2	8	18	18	2	
Hg	80	2	8	18	32	18	2

Chemical reactivity. All three elements can lose the two electrons in the outermost shell to form divalent (with two charges) positive ions, thereby exposing the next innermost shell with a stable configuration in each case of 18 electrons. Ordinary chemical reactions cannot supply enough energy to remove more than two electrons and thus increase the valency above 2, though any number of electrons can be removed under conditions that can provide the necessary energy, such as intense heat or powerful electric or magnetic fields. All three elements tend to use the two outer electrons for covalent bonding; this tendency is most marked in the case of mercury, less so in that of zinc, and least with cadmium.

Zinc is divalent only. It can give up two electrons to form an electrovalent compound; *e.g.,* zinc carbonate $ZnCO_3$. It may also share those electrons, as in zinc chloride, $ZnCl_2$, a compound in which the bonds are partly ionic and partly covalent. Mercury also forms covalent bonds in mercury(II) chloride, $HgCl_2$.

Cadmium compounds are mainly ionic, but cadmium also forms complex ions with ligands (atoms, ions, or molecules that donate electrons to a central metal ion); *e.g.,* the complex ion with ammonia NH_3, having the formula $[Cd(NH_3)_4]^{2+}$, or with the cyanide ion, formula

Table 15: Some Properties of the Zinc Group Elements

	zinc	cadmium	mercury
Atomic number	30	48	80
Atomic weight	65.37	112.41	200.61
Melting point (°C)	419.57	321	−38.8
Boiling point (°C)	907	767	357
Density, solid (g/cm³)	7.14 (20° C)	8.65 (20° C)	14.17 (−38.9° C)
Density, liquid (g/ml)	6.48 (500° C)	7.821 (500° C)	13.546 (20° C)
Valence	2	2	2, 1
Electron configuration	$(Ar)3d^{10}\,4s^2$	$(Kr)4d^{10}\,5s^2$	$(Xe)4f^{14}\,5d^{10}\,6s^2$
Isotopic abundance (terrestrial, percent)	^{64}Zn (50.9), ^{66}Zn (27.3), ^{67}Zn (3.9), ^{68}Zn (17.4), ^{70}Zn (0.5)	^{106}Cd (1.4), ^{108}Cd (1.0), ^{110}Cd (12.8), ^{111}Cd (13.0), ^{112}Cd (24.2), ^{113}Cd (12.3), ^{114}Cd (28.0), ^{116}Cd (7.3)	^{196}Hg (0.15), ^{198}Hg (10.1), ^{199}Hg (17.0), ^{200}Hg (23.3), ^{201}Hg (13.2), ^{202}Hg (29.5), ^{204}Hg (6.7)
Radioactive isotopes (mass numbers)	60–63, 65, 69, 71, 72	103–105, 107, 109, 115, 117–119, 121	185–195, 197, 203, 205, 206
Heat of fusion (cal/g)	24.1	13.2	2.82
Heat of vaporization (cal/g)	426	240	70.6
Specific heat (cal/g/°C at 20° C)	0.092	0.055	0.033
Electrical resistivity at 20° C (micro-ohm-cm)	5.75	7.6	95.8
Hardness (Brinell number)	38.1	29.0	—
Crystal structure	hexagonal, close-packed	hexagonal, close-packed	rhombohedral
Radius			
Metallic (Å)	1.332	1.489	1.500
Ionic (M^{2+}, Å)	0.75	0.99	1.12
Ionization energy (electron volt)			
First	9.36	8.96	10.2
Second	17.89	16.84	18.65
Third	40.0	38.0	—
Electronegativity (Pauling)	1.6	1.7	1.9

$[Cd(CN)_4]^{2-}$. Differing from zinc and mercury, cadmium can form the complex ions represented by the formulas $(CdCl_3)^-$ and $(CdCl_4)^{2-}$ in solution.

Mercury forms the mercury(II) ion Hg^{2+} and the mercury(I) ion $(Hg_2)^{2+}$. In the latter, two electrons are shared in a covalent bond between the two metal atoms. The mercury(I) ion shows little tendency to form complexes, whereas the mercury(II) ion does form them. In contrast to mercury(II) compounds, which are usually covalent, all the common mercury(I) salts are ionic, and the soluble compounds—*e.g.,* mercury(I) nitrate, $Hg_2(NO_3)_2$—show normal properties of ionic compounds, such as ease of dissociation or breakup into separate ions.

Jnique jualities of nercury

Mercury is exceptional in that, unlike zinc or cadmium, it does not react easily with oxygen on heating, and mercury(II) oxide does not show the acid property of forming salts (mercurates), whereas zinc oxide does this readily. Mercury is again anomalous in that it does not produce hydrogen, as do zinc and cadmium, upon treatment with dilute acids. With fairly concentrated nitric acid, zinc and cadmium evolve oxides of nitrogen and form zinc or cadmium nitrates; mercury gives both mercury(II) nitrate, $Hg(NO_3)_2$, and mercury(I) nitrate, $Hg_2(NO_3)_2$. A further characteristic of mercury that is uncommon among metals is its readiness to form stable compounds containing a mercury–carbon bond or a mercury–nitrogen bond. As a result, mercury forms a wide variety of organic compounds (compounds that always contain carbon, usually also hydrogen, and often one or more of the elements oxygen, nitrogen, sulfur). On the whole, therefore, the zinc group elements do not show a smooth gradation of properties, mainly because of the number of anomalous properties of mercury, which in many respects shows a greater similarity to silver than to zinc and cadmium.

Analysis. The classical chemical methods of analysis are now rarely employed except for standardization. When this is required, the methods most commonly employed are the titration of zinc (*i.e.,* addition of a measured volume of a standardized solution of ferrocyanide ion until the exact amount necessary for complete reaction has been added), the conversion of cadmium to cadmium sulfide, which is isolated and weighed, and the colorimetric estimation of mercury (comparison of the intensity of the colour produced by reaction with the substance dithizone with that produced by the same treatment of known amounts of mercury). In daily practice, colorimetry and polarography (a method based on the response of electric current to a steadily increasing electromotive force applied to a solution) are widely used but are being rapidly replaced by other techniques of greater rapidity, simplicity, or accuracy. These modern procedures include atomic absorption spectroscopy (based on the absorption of light of certain wavelengths by atoms present in a flame) and X-ray fluorescence (based on the emission of radiation of characteristic wavelengths when X-rays impinge on a sample).

Toxicity of the elements. The toxicity of the metals increases sharply in the order zinc, cadmium, mercury. The toxicity of zinc is low. In drinking water zinc can be detected by taste only when it reaches a concentration of 15 parts per million (ppm); water containing 40 parts per million zinc has a definite metallic taste. Vomiting is induced when the zinc content exceeds 800 parts per million. Cases of fatal poisoning have resulted through the ingestion of zinc chloride or sulfide, but these are rare. Both zinc and zinc salts are well tolerated by the human skin. Excessive inhalation of zinc compounds can cause such toxic manifestations as fever, excessive salivation, and a cough that may cause vomiting; but the effects are not permanent.

Compared with those of zinc, the toxic hazards of cadmium are quite high. It is soluble in the organic acids found in food and forms salts that are converted into cadmium chloride by the gastric juices. Even small quantities can cause poisoning, with the symptoms of increased salivation, persistent vomiting, abdominal pain, and diarrhea. Fatal cases have been reported. Cadmium has its most serious effect as a respiratory poison: a number of fatalities have resulted from breathing the fumes or dusts that arise when cadmium is heated. Symptoms are difficult or laboured breathing, a severe cough, and violent gastrointestinal disturbance.

Mercury and its compounds are highly toxic. They can be handled safely, but stringent precautions must be taken to prevent absorption by inhalation, by ingestion, and through the skin. The main result of acute poisoning is damage to kidneys.

Mercury poisoning

Numerous cases of poisoning through the industrial use of inorganic mercury compounds have been known. In the 19th century the use of mercuric nitrate in the hat industry to carrot, or lay, the felt caused tremors and a physical disturbance that gave rise to the phrase "as mad as a hatter" and consequently was banned. Organic compounds of mercury, most notably the compounds of the aryl and alkyl families, were once widely used, primarily as fungicides in seeds, paint, and paper. The toxicity of such compounds is different. The behaviour of aryl salts—as for example phenylmercuric acetate—in the body is similar to that of inorganic compounds. Both groups if ingested cause vomiting, colic, and diarrhea, and both are skin irritants. No fatal case of aryl salt poisoning has been reported; however, exposure to alkyl salts has caused a number of deaths. The main target seems to be the central nervous system, and alkyl salts are capable of penetrating brain cells. They are only slowly excreted. Concern has been expressed at an apparent buildup of mercury in tuna, swordfish, and salmon, and many countries have set limits on the amounts allowable in edible fish. The use of mercurial fungicides and pesticides and the discharge of mercury-containing industrial wastes were prohibited in the United States in the early 1970s because they were found to cause such contamination.

IMPORTANT APPLICATIONS

Zinc alloys

Zinc. Zinc, cadmium, and mercury are all of considerable commercial importance. The main application of zinc is as a coating for the protection of steel against corrosion. Zinc itself forms an impervious coating of its oxide on exposure to the atmosphere, and, hence, the metal is more resistant to ordinary atmospheres than iron, and it corrodes at a much lower rate. In addition, because zinc tends to oxidize in preference to iron, some protection is afforded the steel surface even if some of it is exposed through cracks. The zinc coating is formed either by alloying or, to a lesser extent, by electroplating. In the alloying or galvanizing process the steel parts, after suitable cleaning, are dipped in a bath of molten zinc. Some alloying occurs at the interface (surface between the two metals), and the zinc coating adheres tenaciously to the steel. In a continuous galvanizing process developed in the 1930s and '40s, which is displacing the batch dipping process for sheet production, steel strip is passed continuously through a cleaning treatment to remove scale and then through a bath of molten zinc. The zinc coating so formed is firmly adherent and can withstand considerable deformation (change of form). In the rolled state, considerable quantities of zinc are used for roofing, particularly in Europe; small additions of copper and titanium improve creep resistance; *i.e.,* resistance to gradual deformation. Alloyed with copper to form brass, zinc has been widely used since Roman times. Another important series of alloys are those formed by the addition of 4 to 5 percent aluminum to zinc; these have a relatively low melting point but possess good mechanical properties and can be cast under pressure in steel dies.

Cadmium. Cadmium is physically similar to zinc but is more dense and soft, takes a high polish, and is resistant to alkalies. Its main application is as an electrodeposited coating. The plated cadmium has a smaller grain size than electro-zinc coatings, and deposits tend to be more uniform and smooth. Consequently, good protection is afforded by thin coatings of cadmium, and thus, in spite of its high price, it is frequently used for the protection of precision parts. Its resistance to marine atmospheres is also superior to that of zinc. An important application of cadmium is its use with either nickel or silver and a caustic potash electrolyte in electrical storage batteries for uses in which lower weight, longer life, and stability upon storage in discharged condition are desirable as in aircraft. The use of cadmium in nuclear reactors depends on its property of absorbing neutrons. Small quantities of cadmium are added to a number of metals to strengthen them; 1 percent added to copper increases its strength and hardness with only a small reduction in electrical conductivity; alloyed with zinc, cadmium forms solders with good shear strength. The use for cadmium compounds in pigments was mentioned earlier.

Mercury. As has been noted, mercury is the only common metal that is liquid at room temperatures; it freezes at −39° C and boils at 357° C. It is a dense, mobile liquid having a lustrous, silvery-white colour. Because it expands rapidly and uniformly as the temperature increases and does not wet glass, it is specially suitable for thermometers. Because it is not attacked by dry air, oxygen, or carbon dioxide at ordinary temperatures and because it has good electrical conductivity, it is widely used in a variety of electrical measuring and control equipment. Lamps containing mercury vapour in a fused silica tube or bulb are used as a source of ultraviolet radiation.

Amalgams of mercury

The capacity of mercury to form amalgams, or liquid alloys, with solid metals is of interest. Gold and silver dissolve readily in mercury, and in the past this property was used in the extraction of these metals from their ores. Copper, tin, and zinc also form amalgams of some importance. An amalgam with silver is used as a filling in dentistry. The use of mercury in the manufacture of chlorine and caustic soda (sodium hydroxide) by electrolysis of brine depends upon the fact that mercury employed as the negative pole, or cathode, dissolves the sodium liberated to form a liquid amalgam. An interesting application, though not of great commercial significance, is the use of mercury vapour instead of steam in some electrical generating plants, the higher boiling point of mercury providing greater efficiency in the heat cycle. (S.W.K.M./Ed.)

Boron group elements

The five elements comprising the so-called boron group are: boron (B), aluminum (Al), gallium (Ga), indium (In), and thallium (Tl). Constituting Group IIIa of the periodic table (see Figure 10), they are characterized as a group by having three electrons in the outermost parts of their atomic structure. Boron, the lightest of these elements, is a nonmetal but the other members of the group are silvery-white metals.

None of these elements was known in a pure state before modern chemistry isolated them. Very soon after a method had been found to produce it in commercial quantities, aluminum revolutionized industry. The other members of the group, including boron, still have little commercial value. Some of the compounds of boron and aluminum, however, are indispensable in modern technology and have been widely used in many parts of the world throughout recorded history.

HISTORY

Isolation of aluminum metal

Historical importance of gallium

The use of a boron compound known as borax (sodium tetraborate, $Na_2B_4O_7 \cdot 10H_2O$) can be traced back to the ancient Egyptians, who used it as a metallurgical flux (a substance that aids the heat joining or soldering of metals), in medicine, and in mummification. During the 13th century, Marco Polo introduced borax into Europe but not until the mid-19th century, when vast deposits of borates were discovered in the Mojave Desert, did borax become relatively common. The ancient Egyptians, Greeks, and Romans used a compound of aluminum known as alum (the compound potassium aluminum sulfate) in dyeing as a mordant; *i.e.,* a substance that fixes dye molecules to the fabric. Lapis lazuli, a rare, dark blue mineral (the compound sodium aluminum silicate containing sulfur), has been widely used as a semiprecious stone throughout history. The metal aluminum was first isolated early in the 19th century but it was not until a modern electrolytic process based on the use of bauxite was developed that commercial production of aluminum became economically feasible. The other elements of the boron group were first detected spectroscopically (*i.e.,* by analysis of the light emitted by or passed through substances containing the element) in the late 19th century. The existence and properties of gallium were predicted by a Russian chemist, Dmitry Ivanovich Mendeleyev, on the basis of the periodic table of the elements that he had developed; the ultimate

Table 16: Some Properties of the Boron Group Elements

	boron	aluminum	gallium	indium	thallium
Atomic number	5	13	31	49	81
Atomic weight	10.811	26.9815	69.72	114.82	204.37
Colour of element	brown	silver white	gray blue	silver white	blue white
Melting point (°C)	2,100–2,200	659.7	29.75	156.4	303
Boiling point (°C)	sublimes 2,550	2,467	2,403	2,000 ± 10	1,457 ± 10
Density					
Solid (g/cm³)	2.34 (20° C)	2.699 (20° C)	5.903 (25° C)	7.31 (20° C)	11.85 (20° C)
Liquid (g/ml)	2.34 (2,100° C)	2.382 (659.7° C)	6.095 (29.8° C)	—	11.29 (306° C)
Valence	3	3	3	3, 1	3, 1
Electronic configuration	$1s^22s^22p^1$	(Ne) $3s^23p^1$	(Ar) $3d^{10}4s^24p^1$	(Kr) $4d^{10}5s^25p^1$	(Xe) $4f^{14}5d^{10}6s^26p^1$
Isotopic abundance (terrestrial, percent)	boron-10 (18.8), boron-11 (81.2)	aluminum-27 (100)	gallium-69 (61.2), gallium-71 (38.8)	indium-113 (4.23), indium-115 (95.77)	thallium-203 (29.52), thallium-205 (70.48)
Radioactive isotopes (mass numbers)	8, 9, 12, 13	24–26, 28–30	63–68, 70, 72–76	106–112, 114, 116–124	191–202, 204, 206–210
Colour imparted to flame	green	colourless	violet	blue	green
Colour of ions in solution					
M^{3+}	—	colourless	colourless	colourless	colourless
M^+	—	—	—	—	colourless
Heat of fusion (kcal per gram atom)	5.3	2.55	1.34	0.78	1.02
Specific heat at 20° C (cal/g/°C)	0.307	0.214	0.079	0.057	0.031
Electrical resistivity (microhm-cm)	7.7×10^{11} (27° C)	2.65 (20° C)	56.8 (20° C)	9 (20° C)	18 (0° C)
Hardness (Mohs' scale)	9.5	2–2.9	1.5–2.5	1.2	1.4 (approx)
Crystal structure at 20° C	α-rhombohedral, β-rhombohedral, tetragonal	face-centred cubic	orthorhombic	face-centred tetragonal	hexagonal close-packed
Radius					
Atomic (Å)	0.80	1.18	1.25	1.50	1.71
Ionic (M^{3+}, Å)	(0.20)	0.50	0.62	0.81	0.95
Ionization energy (electron volts)					
First	8.296	5.984	6.00	5.79	6.106
Second	23.98	18.82	18.9	18.86	20.42
Third	37.92	28.44	30.70	28.0	29.8
Fourth	259.30	119.96	64.2	57.8	50
Oxidation potential at 25° C ($M \rightarrow M^{3+} + 3e^-$, volts)	—	1.67	0.52	0.34	−1.25
Electronegativity (Pauling)	2.0	1.5	1.6	1.7	1.8

discovery of gallium and the accuracy of his description of the properties of the then unknown element convinced scientists of the theoretical soundness of the table. Gallium is one of two metals (the other is cesium) whose melting points are low enough for them to turn to liquid when held in the hand.

GENERAL PROPERTIES OF THE GROUP

Table 16 gives a list of the electronic configurations and several ionization energies of the boron group elements. Every element in the boron group has three electrons in its outermost shell, and for each element there is a sharp jump in the amount of energy required to remove the fourth electron, reflecting the fact that this electron must be removed from an inner shell. Consequently the elements of the group will have maximum oxidation numbers of three, corresponding to loss of the first three electrons, and will form ions with three positive charges.

The apparently erratic way in which ionization energies vary among the elements of the group is due to the presence of the filled inner *d* orbitals in gallium, indium, and thallium, and the *f* orbital in thallium, which do not shield the outermost electrons from the pull of the nuclear charge as efficiently as do the inner *s* and *p* electrons. In Groups Ia and IIa, in contrast to the boron group, outer shell (always referred to as n) electrons are shielded in every case by a constant inner set of electrons, in the (n-1)s^2(n-1)p^6 orbitals, and the ionization energies of these Group-Ia and Group-IIa elements decrease smoothly down the group. The ionization energies of Ga, In, and Tl are thus higher than expected from their Group II counterparts because their outer electrons, being poorly shielded by the inner *d* and *f* electrons, are more strongly bound to the nucleus. This shielding effect also makes the atoms of gallium, indium, and thallium smaller than the atoms of their Group Ia and IIa neighbours by causing the outer electrons to be pulled closer toward the nucleus.

The M^{3+} state for Ga, In, and Tl will be energetically less favourable than Al^{3+} because the high ionization energies of these three elements cannot always be balanced by the crystal energies of possible reaction products. For example, of the simple, anhydrous compounds of trivalent thallium only the trifluoride, TlF_3, is ionic. For the group as a whole, therefore, the M^{3+} ionic state is the exception rather than the rule. More commonly the elements of the group form covalent bonds and achieve an oxidation state of three by promoting one electron from the *s* orbital in the outer shell (designated n*s* orbital) to an n*p* orbital, the shift permitting the formation of hybrid, or combination, orbitals (of the variety designated as sp^2). Increasingly down the group there is a tendency toward the formation of M^+ ions and at thallium the monovalent state is the more stable one. The basicity (a property of metals) of the elements also increases in proceeding down the group as shown by the oxides they form: boric oxide (formula B_2O_3) is acidic, the next three oxides, of aluminum, gallium, and indium (formulas Al_2O_3, Ga_2O_3, and In_2O_3) are either acidic or basic depending on the environment (a property called amphoterism), and thallic oxide (Tl_2O_3) is wholly basic.

The ionization energies listed in Table 16 suggest that the formation of salts of the M^{2+} ions might be feasible. At first glance, such appears to be the case, since gallium compounds with the formula GaX_2 (X representing chlorine, bromine, or iodine) can be made, and similar cases occur with the other metals of this group. Such compounds, however, are generally found to be of mixed valency—that is, they contain metal atoms in both the one and the three oxidation states, a condition symbolized as $M^+(M^{III}X_4)^-$. The nearest approach to M^{2+} derivatives occurs in gallium sulfide, selenide, and telluride, which are made by heating gallium with stoichiometric amounts of sulfur, selenium, and tellurium, respectively. Studies of the structure of these compounds by X-ray methods show that they contain $(Ga\text{-}Ga)^{4+}$ units arranged in a layer-like lattice; the coupling of the gallium atoms in such a manner pairs the electrons available for the bonds and thereby explains the diamagnetism of the compounds (diamagnetism is a property associated with paired electrons).

The large amount of energy required to remove three electrons from a boron atom makes the formation of salts containing the bare B^{3+} cation impossible; even water of hydration associated with such ions would be too highly deformed to be stable and hence the aquated ion $B^{3+}{}_{(aq)}$ is unknown. Much less energy is required to promote electrons from 2*s* orbitals into 2*p* orbitals in boron atoms with the result that boron compounds are always covalent. The boron orbitals are hybridized to either the sp^2 (when boron forms bonds with three other atoms) or the sp^3 (when boron forms bonds with four atoms) configuration (see MOLECULES).

Nuclear resonance studies

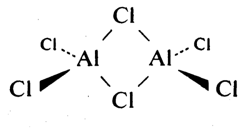
Although simple M^{3+} cations are uncommon in anhydrous compounds of the boron group elements, the hydrated (combined with water) trivalent ions of aluminum, gallium, indium, and thallium are well known in water solution. Nuclear magnetic resonance studies reveal that there are six water molecules held strongly by these positive ions in solution, and their salts often can be crystallized from solution combined with six water molecules. The high charge on the central cation of such hydrates induces the ionization of protons, or hydrogen nuclei, on the coordinated water molecules and thereby leads to the formation of basic salts. This reaction is represented in the following equations:

$$M(H_2O)_n^{3+} \rightleftharpoons (HO)M(H_2O)_{n-1}^{2+} + H_{aq}^+$$
$$\rightleftharpoons (HO)_2M(H_2O)_{n-2}^{+} + 2H_{aq}^+ \rightleftharpoons M(OH)_3 + 3H_{(aq)}^+$$

in which, as before, M represents an ion of one of the boron group elements; *n* is the number of water molecules joined to it; (HO)M represents a hydroxide group joined to the metal ion; and $H(aq)^+$ is a hydrated hydrogen ion. In these and other equations the arrows pointing in two directions indicate that the chemical reactions can proceed both ways depending on the reaction conditions. When acid is added to such aqueous solutions it depresses the hydrolytic processes by reversing the above reactions. At high acid concentrations, however, complex anions (negative ions) are sometimes formed, especially with the aqueous hydrogen halides. The following equation illustrates this: $Ga(aq)^{3+} + HX\ (conc.) \rightarrow GaX_4^-$, X being chlorine, bromine, or iodine. Intermediate complex ions, MX^{2+} and MX_2^+ can be detected in several cases.

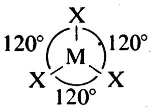
Conductivity of solid aluminum trichloride

The electrical conductivity of solid aluminum trichloride (formula $AlCl_3$), in which each aluminum ion has three positive charges, increases rapidly as the temperature is elevated toward the melting point, at which the conductivity suddenly falls to zero. This phenomenon occurs because the aluminum and chloride ions form an ionic lattice that partially conducts electricity; but upon melting, the compound changes to the electrically nonconducting, covalent state. The explanation is that the distribution of energy in the liquid state is insufficient to compensate for the ionization energy required to separate the Al^{3+} and Cl^- ions and these then acquire covalent bonds. The liquid consists of double or dimeric molecules with the formula Al_2Cl_6, which may be represented in the following manner that shows a molecule with the position of its atoms in three dimensions; the solid lines are in the plane of the paper, the dotted lines are behind the paper, and the shaded lines indicate that they extend toward the viewer:

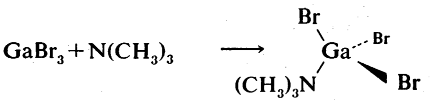


The delicate energy balance between ionic and covalent bonding for aluminum in the trivalent state can be appreciated when it is realized that whereas solid aluminum trifluoride, formula AlF_3, is ionic like the chloride, aluminum tribromide forms molecular crystals containing dimers, with the formula Al_2Br_6.

In contrast with the dimers, the single, or monomeric, trihalides of the boron group elements have trigonal planar structures. If M is the metal and X is any halogen, the arrangement of the atoms can be sketched as follows:

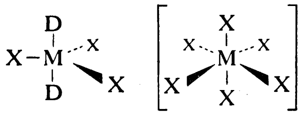


The trihalides of boron have this configuration in all phases whereas the trihalides of Al, Ga, In, and Tl become monomeric only on being heated in the gas phase. In MX_3 molecules, the central atom M has added three electrons to its own making only six electrons in the outer shell, although eight are required to achieve the desired inert-gas configuration. These halides, therefore, readily accept two more electrons from many donor molecules (*e.g.*, ethers, alcohols, amines, and phosphines) that carry unshared pairs of electrons. A typical case, the reaction of gallium tribromide with trimethylamine, is represented in the following equation:



The central gallium atom is coordinated or bonded to three bromine atoms and one nitrogen atom. The electron donor also can be a halide ion, in which case the tetrahedral complex anion, MX_4^- results.

A few compounds are known in which aluminum, gallium, indium, and thallium are coordinated to five or six atoms. These compounds have structures of the following types, M again representing any boron group element, D any donor molecule, and X any halogen (again, the solid lines are bonds in the plane of the paper, the atoms so bonded lying in that plane; the dotted lines lead behind the paper; the shaded lines reach toward the viewer):



In such compounds it is possible, but by no means certain, that the central element makes use of its vacant n*d* orbitals (see above) to increase its valency by way of sp^3d (five-coordination) or sp^3d^2 (six-coordination) hybridization. If the concept of the participation of *d* orbitals in the bonding of these compounds is valid, it would account for the fact that boron, which has no available *d* orbitals, does not form five- and six-coordinate compounds. In many cases, however, spatial requirements also would rule out the possibility of boron increasing its covalency above four because the boron atom is so small no more than four atoms can be arranged around it.

In the gas phase at high temperature all the boron group elements form diatomic halides MX, either by dissociation of the trihalides or, more commonly, by reduction of the trihalides with the free element as in the following equations for two such reactions:

$$\underset{\text{boron}}{2B} + \underset{\text{boron trifluoride}}{BF_3} \xrightarrow{2{,}000°} \underset{\text{boron monofluoride}}{3BF}$$

$$\underset{\text{aluminum}}{2Al} + \underset{\text{aluminum trichloride}}{AlCl_3} \xrightarrow{800°-1{,}000°} \underset{\text{aluminum monochloride}}{3AlCl}.$$

Most of these monohalides, especially those of boron, aluminum, and gallium, are unstable in the solid state under normal conditions; they exist only at high temperatures as gases; all are covalently bonded, except thallium fluoride, which exists as the ion pair, Tl^+F^-.

Thallium is the only element that forms a stable ion having an $(n\text{-}1)d^{10}ns^2$ outer electronic configuration. There is, therefore, no ion to which direct comparisons with the thallium(I) ion, Tl^+, might be made.

INDIVIDUAL BORON GROUP ELEMENTS

Boron. The first three ionization energies of boron, atomic number 5, are much too high to allow the formation of compounds containing the B^{3+} ion; in all its compounds boron is covalently bonded. One of boron's 2*s* electrons is promoted to a 2*p* orbital, giving the outer electron configuration $2s^12p^2$; the *s* and *p* orbitals can then be mixed to give sp^2 and sp^3 hybrids, which allow boron to be three- and four-coordinated, respectively. The three-coordinate derivatives (*e.g.*, halides, alkyls, aryls) are planar molecules that readily form donor-acceptor complexes (called adducts) with compounds containing lone pairs of electrons; in these adducts the boron atom is four coordinated, the four groups being tetrahedrally disposed around it.

Biological importance

Boron is an essential trace element for the healthy growth of many plants, and typical effects of long-term boron deficiency are stunted, misshapen growth and, in root crops, heart rot and dry rot; the deficiency can be alleviated by the application of soluble borates to the soil. Gigantism of several species of plants growing in soil naturally abundant in boron has been reported. It is not yet clear what the precise role of boron in plant life is, but most researchers agree that the element is in some way essential for the normal growth and functioning of apical meristems, the growing tips of plant shoots.

Boron is unique in its group in that it forms a rather large number of compounds with hydrogen, or hydrides, called boranes, many of which have most unusual, three-dimensional structures; it also forms a series of halides with the general formula B_nX_n. These molecules are interesting because they contain closed cages of boron atoms. Examples are the boron chlorides whose formulas are B_4Cl_4, B_8Cl_8, and B_9Cl_9. Unfortunately these interesting halides, most of which are highly coloured in sharp contrast to the more typical boron derivatives, are exceedingly difficult to prepare and to handle. The substance, formula B_4Cl_4, for example, can be prepared only in milligram quantities, and complex electrical-discharge techniques are needed for its production; furthermore, it ignites spontaneously in air and is rapidly decomposed both by water and even by the grease used to lubricate the vacuum equipment employed in its preparation. Closed cages containing 12 boron atoms arranged in the form of an icosahedron (see Figure 12) also occur in the various crystalline forms of elemental boron.

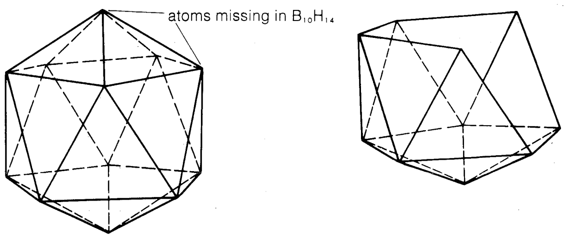


Figure 12: (Left) Icosahedron, basic unit of crystalline boron. (Right) $B_{10}H_{14}$ skeleton (see text).

Boron reacts with all halogen elements to give monomeric, highly reactive trihalides, which readily form complexes with amines, phosphines, ethers, and halide ions. Examples of complex formation between boron trichloride and trimethylamine, as well as between boron trifluoride and fluoride ion, are shown in the following equations:

$$BCl_3 + N(CH_3)_3 \longrightarrow (CH_3)_3N \cdot BCl_3$$

$$BF_3 + F^- \longrightarrow BF_4^-$$

in which the heavy dot indicates that a bond is formed between the nitrogen and boron atoms. When boron trichloride is passed at low pressure through devices delivering an electric discharge, diboron tetrachloride, the formula written as $Cl_2B\text{–}BCl_2$, and tetraboron tetrachloride, formula B_4Cl_4, are formed. Diboron tetrachloride decomposes at room temperatures to give a series of monochlorides having the general formula $(BCl)_n$, in which *n* may be 8, 9, 10, or 11; the compounds with formulas B_8Cl_8 and B_9Cl_9 are known to contain closed cages of boron atoms.

Nuclear properties of boron

Boron exists in nature as two isotopes, one of atomic mass 10 (18.8 percent) and one of atomic mass 11 (81.2 percent). Both nuclei possess nuclear spin (rotation of the atomic nuclei); that of boron-10 has a value of 3 and that of boron-11, 3/2, the values being dictated by quantum factors. These isotopes are, therefore, of use in nuclear magnetic resonance spectroscopy; and spectrometers specially adapted to detecting the boron-11 nucleus are available commercially. The boron-10 and boron-11 nuclei also cause splitting in the resonances (that is, the appearance of new bands in the resonance spectra) of other nuclei (*e.g.*, those of hydrogen atoms bonded to boron). The boron-10 isotope is unique in that it possesses an extremely large capture cross section for thermal neutrons (*i.e.*, it readily absorbs neutrons of low energy). The capture of a neutron by a nucleus of this isotope results in the expulsion of an alpha particle (nucleus of a helium atom, symbolized α):

$$^{10}\mathrm{B} + {}^{1}n \longrightarrow {}^{7}\mathrm{Li} + {}^{4}a.$$

Since the high-energy alpha particle does not travel far in normal matter, boron may be used in the fabrication of neutron shields (materials not penetrable by neutrons). In the Geiger counter, alpha particles trigger a response, whereas neutrons do not; hence if the gas chamber of a Geiger counter is filled with a gaseous boron derivative (*e.g.*, boron trifluoride), the counter will record each alpha particle produced when a neutron that passes into the chamber is captured by a boron-10 nucleus. In this way, the Geiger counter is converted into a device for detecting neutrons, which normally do not affect it. The affinity of boron-10 for neutrons also forms the basis of a technique for treating patients suffering from brain tumours. For a short time after certain boron compounds are injected into a patient with a brain tumour, the compounds collect preferentially in the tumour; irradiation of the tumour area with thermal neutrons, which cause relatively little general injury to tissue, results in the release of a tissue-damaging alpha particle in the tumour each time a boron-10 nucleus captures a neutron. In this way destruction can be limited preferentially to the tumour, leaving the normal brain tissue less affected.

The presence of boron compounds can be detected qualitatively by the green coloration they impart to the flame of an ordinary laboratory, or bunsen, burner. Quantitatively, boron is most easily analyzed by converting the material to be analyzed into boric acid by treatment with acid; the excess mineral acid is then neutralized and the much weaker boric acid is titrated (neutralized on a volume-volume basis) in the presence of a sugar, such as mannitol, to make the acid detectable.

Aluminum. In aluminum, atomic number 13, the configuration of the three outer electrons is such that in a few compounds (*e.g.*, crystalline aluminum fluoride [AlF_3] and aluminum chloride [$AlCl_3$]) the bare ion, Al^{3+}, formed by loss of these electrons, is known to occur. The energy required to form the Al^{3+} ion, however, is very high; and, in the majority of cases, it is energetically more favourable for the aluminum atom to form covalent compounds by way of sp^2 hybridization, as boron does. The Al^{3+} ion can be stabilized by hydration, and the octahedral ion $Al(H_2O)_6{}^{3+}$ occurs both in aqueous solution and in several salts. The alums, double salts of formula $M^{I}Al(SO_4)_2 \cdot 12H_2O$, also contain this ion; the univalent ion M can be sodium, potassium, rubidium, cesium, ammonium, or thallium, and the aluminum may be replaced by a variety of other M^{3+} ions; *e.g.*, gallium, indium, titanium, vanadium, chromium, manganese, iron, or cobalt. The name aluminum (British usage is aluminium) is derived from the Latin word *alumen* used to describe potassium alum, formula $KAl(SO_4)_2 \cdot 12H_2O$.

Aluminum can be detected in concentrations as low as one part per million using emission spectroscopy. Aluminum can be quantitatively analyzed as the oxide (formula Al_2O_3) or as a derivative of the organic nitrogen compound 8-hydroxyquinoline. The derivative has the molecular formula $Al(C_9H_6ON)_3$.

Gallium. Gallium, atomic number 31, is comparable to aluminum in its chemical properties. It does not dissolve in nitric acid because a protective film of gallium oxide is formed over the surface by the action of the acid, but the metal does dissolve in other acids to give gallium salts and it dissolves in alkalies, with evolution of hydrogen, to give gallates (compounds in which gallium appears in the anion). Of the halides, only gallium trifluoride is ionic; the others have molecular lattices containing dimeric molecules, with formula Ga_2X_6. Unlike aluminum, however, gallium forms several derivatives that contain gallium(I) oxidation state; for example, the oxide, formula Ga_2O. The sulfide (GaS), selenide (GaSe), and telluride (GaTe), made directly by combination of the elements at high temperature, are diamagnetic and contain gallium—gallium units with four positive charges $(Ga—Ga)^{4+}$, in a layer lattice. The metal gallium is stable in dry air. On burning in air or oxygen it forms the white oxide, formula Ga_2O_3. This oxide can be reduced to the metal when heated strongly in hydrogen, and with gallium metal at 700° C it gives the lower oxide Ga_2O. The hydroxide, formula $Ga(OH)_3$, is amphoteric; *i.e.*, it reacts either as an acid or a base depending on the circumstance; it is precipitated from solutions of gallium salts by alkali hydroxides.

Indium. Indium, atomic number 49, is an amphoteric element; it dissolves in acids to give indium salts and it also dissolves in concentrated alkalies to give indates. All anhydrous indium(III) derivatives, except indium trifluoride, formula InF_3, are covalent. There is a marked tendency for two of the outer electrons of the indium atom (the outer $5s^2$ electrons) not to be used in bonding; this circumstance results in univalent indium compounds.

Indium burns to the yellow oxide, formula In_2O_3, when heated in air or oxygen; a better method of preparation for the oxide is to heat the hydroxide, nitrate, or sulfate. This oxide is easily reduced to the metal, and on strong heating it loses oxygen to give the monoxide In^{I}_2O. Indium hydroxide dissolves in both acids and alkalies.

Monovalent thallium

Thallium. Thallium, atomic number 81, is typical of the Group IIIa elements in having an s^2p^1 outer electron configuration. Promoting an electron from an s to a p orbital allows the element to be three or four covalent. With thallium, however, the energy required for $s \rightarrow p$ promotion is high relative to the Tl–X covalent bond energy which is regained on formation of TlX_3; hence, a trivalent derivative is not a very energetically favoured reaction product. Thallium normally forms the more stable thallium (I), or thallous, salts containing the ion Tl^+ (in which the $6s^2$ electrons remain unused); it is the only element to form a stable univalent cation with the outer electron configuration $(n\text{-}1)d^{10}ns^2$, which is, unusually enough, not an inert gas configuration. In its oxidation state three, indicated as thallium(III), or commonly called thallic, thallium resembles aluminum, although the ion Tl^{3+} appears to be too large to form alums. In the monovalent state, thallium(I) displays properties similar to those of both the heavier alkali metals and silver. The very close similarity in size of the thallium(I) ion, Tl^+, and the rubidium ion, Rb^+, makes many thallium(I) salts, such as the chromate, sulfate, nitrate, and halides, isomorphous (*i.e.*, have an identical crystal structure) to the corresponding rubidium salts; also the ion Tl^+ is able to replace the ion Rb^+ in the alums. Thus, thallium does form an alum; but in doing so it replaces the M^{I} ion, rather than the expected metal atom M^{III}, in the formula $M^{I}M^{III}(SO_4)_2 \cdot 12H_2O$.

Thallium imparts a brilliant green coloration to a bunsen flame. Thallous chromate, formula Tl_2CrO_4, is best used in the quantitative analysis of thallium, after any thallic ion present in the sample has been reduced to the thallous state. (A.G.M.)

Carbon group elements

The carbon group of five chemical elements consists of carbon (C), silicon (Si), germanium (Ge), tin (Sn), and lead (Pb); they comprise Group IVa in the periodic table (see Figure 10). Except for germanium, all of these elements are familiar in daily life either as the pure element or in the form of compounds, although, except for silicon, none is particularly plentiful in the Earth's crust. Carbon forms an almost infinite variety of compounds, in both the plant and animal kingdoms. Silicon and silicate minerals

Occurrence

are fundamental components of the Earth's crust; silica (silicon dioxide) is sand. Tin and lead, with abundances in the crust lower than those of some so-called rare elements, are nevertheless common in everyday life. They occur in highly concentrated mineral deposits, can be obtained easily in the metallic state from those minerals, and are useful as metals and as alloys in many applications. Germanium, on the other hand, forms few characteristic minerals and is most commonly found only in small concentrations in association with the mineral zinc blende and in coals. Although germanium is indeed one of the rarer elements, it assumed importance upon recognition of its properties as a semiconductor (*i.e.,* limited ability to conduct electricity).

HISTORY

Carbon as an element was discovered by the first man to handle charcoal from his fire; thus, together with sulfur, iron, tin, lead, copper, mercury, silver, and gold, carbon was one of the small group of elements well known in the ancient world. Modern carbon chemistry dates from the development of coals, petroleum, and natural gas as fuels and from the elucidation of synthetic organic chemistry, both substantially developed since the 1800s.

Amorphous elemental silicon was first obtained in a state of purity in 1824 by the Swedish chemist Jöns Jacob Berzelius; impure silicon had already been obtained in 1811. Crystalline elemental silicon was not prepared until 1854, when it was obtained as a product of electrolysis. In the form of rock crystal, however, silicon was familiar to the predynastic Egyptians, who used it for beads and small vases; to the early Chinese; and probably to many others of the ancients. The manufacture of glass containing silica was carried out both by the Egyptians—at least as early as 1500 BC—and by the Phoenicians. Certainly, many of the naturally occurring compounds called silicates were used in various kinds of mortar for construction of dwellings by the earliest men.

Mendeleyev's prediction

Germanium is one of three elements the existence of which was predicted in 1871 by the Russian chemist Dmitry Ivanovich Mendeleyev when he first devised his periodic table. Not until 1886, however, was germanium identified as one of the elements in a newly found mineral.

The origins of tin also are lost in antiquity. Apparently, bronzes, which are copper–tin alloys, were used by man in prehistory long before pure tin metal itself was isolated. Bronzes were common in early Mesopotamia, the Indus Valley, Egypt, Crete, Israel, and Peru. Much of the tin used by the early Mediterranean peoples apparently came from the Scilly Islands and from Cornwall in the British Isles, where tin mining dates to at least 300–200 BC. Tin mines were operating in both the Inca and Aztec domains of South and Central America before the Spanish conquest.

Lead is mentioned often in early Biblical accounts. The Babylonians used the metal as plates on which to record inscriptions. The Romans used it for tablets, water pipes, coins, and even cooking utensils; indeed, as a result of the last use, lead poisoning was recognized in the time of Augustus Caesar. The compound known as white lead was apparently prepared as a decorative pigment at least as early as 200 BC. Modern developments date to the exploitation in the late 1700s of deposits in the Missouri–Kansas–Oklahoma area in the United States.

COMPARATIVE CHEMISTRY

In the periodic table, the elements with eight electrons outermost form the group known as the noble gases (Group 0), the least reactive of the elements. The carbon group elements (Group IVa), with four electrons, occupy a middle position. Elements to the left of Group IVa have fewer than four electrons in the valence shell and tend to lose them (with their negative charges) to become positively charged ions, represented by the symbol for the element with a superscript indicating the number and sign of the charges; such elements are called metals. The nonmetals (except boron) are in the groups to the right of Group IVa; each has more than four electrons in its outermost shell and tends to acquire electrons to complete its octet, forming negatively charged ions.

Chemical reactions result from the exchange of electrons among atoms. In general, if a metal loses its few valence electrons to a nonmetal, the resulting oppositely charged ions are attracted to one another and form a bond, classified as ionic or electrovalent. Two nonmetals, neither of which can actually lose its valence electrons in chemical reaction, may nevertheless share them in pairs in such a way that what is called a covalent bond results. Metal atoms will bond to one another in a third type of bond, which releases their valence electrons in a way that allows them to conduct electricity.

Formation of covalent bonds with non-metals

All the carbon group atoms, having four valence electrons, form covalent bonds with nonmetal atoms; carbon and silicon cannot lose or gain electrons to form free ions, whereas germanium, tin, and lead do form metallic ions but only with two positive charges. Even lead, the most metallic of the carbon group atoms, cannot actually lose all four of its valence electrons, because, as each one is removed, the remainder are held more strongly by the increased positive charge. Because the distinction between covalent and electrovalent (ionic) bonds is often a matter of convenience for the chemist, and because the actual bond structure within a molecule may be quite complicated, it is often useful instead simply to count the total number of electrons an element gains or loses in bonding without regard to the nature of the bonds. This number is called the oxidation number, or oxidation state, of the element; many elements have more than one oxidation state possible, each oxidation state being found in different compounds. The oxidation state of an element is written as a Roman numeral following the name of the element in a compound—*e.g.,* lead(II) means lead in oxidation state plus two; with the chemical symbol of the element, the oxidation state may be written as a superscript, as in Pb^{II}. Covalent bonds generally are considered to be formed by interaction of the orbitals (generally, only the *s, p,* and *d* orbitals) in specific and varied ways. The most common are called sigma and pi bonds, written σ and π, respectively. The sigma bonds are symmetrical with respect to the axis of the bond, whereas the pi bonds are not. Examples of sigma and pi bonding as well as ionic bonding are found among the compounds of the elements of the carbon group. (For a full general discussion and further information useful to an understanding of the following sections, see ATOMS; MOLECULES; and CHEMICAL REACTIONS.)

GENERAL PROPERTIES OF THE GROUP

Metals and nonmetals of the group

The properties of the carbon group elements and those of their compounds are intermediate between properties associated with the elements of the adjacent boron and nitrogen groups. In all groups the metallic properties, resulting from the tendency to hold valence electrons more loosely, increase with atomic number. Within the carbon group, more than in any other, the change from nonmetallic to metallic character with increasing atomic number is particularly apparent. Carbon is a true nonmetal in every sense. Lead is a true metal. Silicon is almost completely nonmetallic; tin is almost completely metallic. Germanium is metallic in appearance and in a number of its other physical properties (see Table 17), but the properties of many of its compounds are those of derivatives of nonmetals. These changes are consequences of increase in atomic size with substantial screening of the larger nuclear charge by intervening electronic shells, as evidenced by decrease in ionization energy (energy required to remove an electron) and electronegativity power to attract electrons with increasing atomic number.

Crystal structure. In the solid state, elemental carbon, silicon, germanium, and gray tin (defined as alpha [α] tin; see Table 17) exist as cubic crystals, based upon a three-dimensional arrangement of bonds. Each atom is covalently bonded to four neighbouring atoms in such a way that they form the corners of a tetrahedron (a solid consisting of four three-sided faces). A practical result is that no discrete small molecules of these elements, such as those formed by nitrogen, phosphorus, or arsenic, can be distinguished; instead, any solid particle or fragment of one of these elements, irrespective of size, is uniformly

Table 17: Some Properties of the Carbon Group Elements

	carbon	silicon	germanium	tin	lead
Atomic number	6	14	32	50	82
Atomic weight	12.011	28.086	72.59	118.69	207.19
Colour of element	colourless (diamond, D), black (graphite, G)	gray	white metallic	white metallic (beta, β), gray (alpha, α)	bluish-white metallic
Melting point (°C)	>3,500 (D)	1,420	937.4	231.9 (α)	327.5
Boiling point (°C)	4,827 (D)	2,355	2,830	2,270 (α)	1,744
Density Liquid (g/ml)				6.97 (231.9° C)	10.51 (400° C)
Solid	3.51 (D), 2.25 (G)	2.33 (25° C)	5.323 (25° C)	5.75 (α), 7.28 (β)	11.35 (20° C)
Oxidation states	−4, (+2), +4	−4, (+2), +4	−4, +2, +4	(−4), +2, +4	(−4), +2, +4
Electronic configuration	(He) $2s^22p^2$	(Ne) $3s^23p^2$	(Ar) $3d^{10}4s^24p^2$	(Kr) $4d^{10}5s^25p^2$	(Xe) $4f^{14}5d^{10}6s^26p^2$
Isotopic abundance (terrestrial, percent)	^{12}C (98.89), ^{13}C (1.11)	^{28}Si (92.21), ^{29}Si (4.70), ^{30}Si (3.09)	^{70}Ge (20.53) ^{72}Ge (27.43) ^{73}Ge (7.76) ^{74}Ge (36.54) ^{76}Ge (7.76)	^{112}Sn (0.96), ^{114}Sn (0.66), ^{115}Sn (0.35), ^{116}Sn (14.3), ^{117}Sn (7.61), ^{118}Sn (24.03), ^{119}Sn (8.58), ^{120}Sn (32.85), ^{122}Sn (4.72), ^{124}Sn (5.94)	^{204}Pb (1.48), ^{206}Pb (23.6), ^{207}Pb (22.6), ^{208}Pb (52.3)
Radioactive isotopes (mass numbers)	10, 11, 14, 15, 16	25–27, 31, 32	65–69, 71, 75, 77, 78	108–111, 113, 121, 123, 125–128	194–203, 205, 209–214
Heat of fusion (kcal/g-atom)		11.1	7.3	1.69	1.21
Heat of vaporization (kcal/g-atom)	170 (G)	71	78	61.8	42.3
Heat of sublimation (kcal/g-atom)	170	85		78	47.5
Heat capacity (cal/mole °C)	1.462 (D), 2.038 (G)	5.086 (0–100° C)	5.372	6.29 (25° C)	6.53 (15–327.5° C)
Critical temperature (°C)		about 4,920			
Critical pressure (atm)		1,450			
Electrical resistivity (microhm-cm)	1,375.0	10		11	20.648
Hardness (Mohs scale)	10 (D) 0.5–1.5 (G)	7	6	1.5–1.8	1.5
Crystal structure	cubic (D) hexagonal (G)	cubic (diamond)	cubic (diamond)	cubic (diamond, α) tetragonal (β)	close-packed, metallic
Radius					
Covalent (Å)	0.772	1.17	1.22	1.40	
Metallic (Å)			1.22	1.40	1.54
Ionic (Å)	2.60 (M^{4-}), 0.16 (M^{4+})	2.71 (M^{4-}), 0.42 (M^{4+})	2.72 (M^{4-}), 0.73 (M^{2+}), 0.53 (M^{4+})	2.94 (M^{4-}), 0.93 (M^{2+}), 0.71 (M^{4+})	1.20 (M^{2+}), 0.84 (M^{4+})
Ionization energy (kcal/g-atom)					
First	261	189	184	171	172
Second	564	378	369	339	348
Third	1,103	773	791	705	738
Fourth	1,486	1,040	1,053	941	(977)
Electronegativity					
(Sanderson)	2.47	1.74	2.31	2.02	2.01
(Pauling)	2.5	1.8	1.8	1.8	1.8

bonded throughout, and, therefore, the whole fragment can be considered as a giant molecule. Decreasing melting points, boiling points, and decreasing heat energies associated with fusion (melting), sublimation (change from solid to gas), and vaporization (change from liquid to gas) among these four elements, with increasing atomic number and atomic size, indicate a parallel weakening of the covalent bonds in this type of structure. The actual or probable arrangement of valence electrons is often impossible to determine, and, instead, relative energy states of the electrons, in the ground, or least energetic, state of the atom are considered. Thus, the same trend of nonmetallic toward metallic states is indicated by decreasing hardness and decreasing single-bond energy between atoms. Carbon crystallizes in two forms, as diamond and as graphite; diamond stands apart from all other elemental forms in the extreme stability of its crystal structure, whereas graphite has a layer structure. As may be expected, cleavage between layers of graphite is much easier to effect than rupture within a layer. The crystal structures of white beta (β) tin and elemental lead are clearly metallic structures. In a metal, the valence electrons are free to move from atom to atom, and they give the metal its electrical conductivity.

Uses. All the carbon group elements and many of their compounds have important uses. Carbon as diamond, for example, is the most expensive and brilliant of all the natural gemstones and also the hardest of abrasives; carbon as graphite is important as electrodes in electrochemical cells, as a lubricant, and, in microcrystalline and nearly amorphous form, as a black pigment, an adsorbent, a fuel, a filler for rubber, and in pencils. Coals are elemental carbon mixed with varying amounts of carbon compounds; coke and charcoal are nearly pure carbon. All organic compounds, such as proteins, carbohydrates, and fats, contain carbon, and all plant and animal cells consist of carbon compounds and their polymers. (Polymers are macromolecules consisting of many simple molecules bonded together in specific ways.) Carbonate minerals are important sources of various metals, such as sodium, magnesium, calcium, copper, and lead.

Elemental silicon is important largely as a semiconductor. Silica (silicon dioxide) is useful as an abrasive, in the production of glass and other ceramic bodies, as an adsorbent, and as sand in mortars and concretes; and these are only a few of its applications. Both naturally occurring and synthetically produced silicates are important in ceramics, building materials, absorbents, and ion exchangers.

Applications of germanium have been limited to semiconducting devices, indicating its weak metallic nature.

Tin-plating of iron protects the latter from corrosion; tin piping and valves maintain purity in water and beverages; molten tin is the base for (float) plateglass production; and tin is a significant component of such alloys as bronzes, pewter, bearing metals, type metals, and lead-based solders. Tin exists in two oxidation states; tin(IV) oxide is useful in making ceramic bodies opaque, as a mild abrasive, and as a weighting agent for fabrics; tin(II) fluoride and tin(II) pyrophosphate are used in dentifrices. Organic tin compounds act as stabilizers in certain plastics and as wood preservatives.

Oxidation states of lead

Elemental lead finds extensive use. Together with the compound lead(IV) oxide and with lead–antimony or lead–calcium alloys, it is employed in common storage batteries. Type metals, bearing alloys, shot metal, and fusible alloys all contain lead. Tetraethyllead is the most important antiknock additive to motor fuels.

Electron configurations. The ground-state electronic configurations (see Table 17) of atoms of these carbon group elements show that each has four electrons in its outermost shells. As has been explained, if *n* represents the outermost shell (*n* being two for carbon, three for silicon, etc.), then these four electrons are represented by the symbols ns^2np^2. Such a configuration suggests the importance of referring to the relatively stable noble-gas-atom configuration preceding each element in determining the properties of the element, in particular its chemical properties. The loss of four electrons by either a carbon atom or a silicon atom to give ions having a positive charge of four (or +4, written C^{4+} or Si^{4+}) with the electron configurations of the preceding noble-gas atoms is precluded by the sizable ionization energies as shown in Table 1. Ions of +4 charge do not exist, nor is there any evidence that carbon or silicon ions of charge +2 can form by the loss of only two unpaired (*np*, or outermost) electrons. Electron loss by atoms of the heavier elements of the family is easier, but it cannot lead to ions with noble-gas-atom configurations because of the presence of underlying (*i.e.*, d^{10}) arrangements of electrons inside the outermost shell. It is again unlikely that the +4 ions of germanium, tin, and lead (in symbols Ge^{4+}, Sn^{4+}, and Pb^{4+}) exist in known compounds, but it is true that the inertness of the ns^2 pair of electrons (which are, in terms of energy states, closer to the nucleus than the np^2 pair) increases substantially with increasing atomic number in the family and thus allows the np^2 electrons to be removed separately, to form at least the ions, Sn^{2+} and Pb^{2+}. Positive two and four oxidation states can be assigned in covalent unions of each of these elements with elements that are more electronegative (*i.e.*, having greater affinity for electrons).

Catenation. Carbon is unique among the elements in the almost infinite capacity of its atoms to bond to each other in long chains, a process called catenation. This characteristic reflects the strength of the bond between adjacent carbon atoms in the molecule, both in relationship to similar bonds involving other elements of the carbon family and in relationship to bonds between carbon atoms and atoms of many other elements. Only the carbon–hydrogen, carbon–fluorine, and carbon–oxygen single bonds (C—H, C—F, and C—O) are stronger than the carbon–carbon single bond (C—C), and each of these is weaker than the carbon–carbon multiple bonds ($C{=}C$ or $C{\equiv}C$). On the other hand, the silicon–silicon single bond (Si—Si) is weaker than other single bonds involving an atom of other elements with the silicon atom. The same is undoubtedly true of the germanium–germanium and tin–tin single bonds (Ge—Ge, Sn—Sn) in relationship to single covalent bonds between atoms of these elements and atoms of other elements. Experimentally, there appears to be no practical upper limit to catenation involving carbon. This phenomenon in three dimensions produces the diamond and in two dimensions the layers in graphite. Catenation is also exhibited to a high degree by elemental silicon, germanium, and tin, but it is strictly limited in compounds of these elements; silicon may have up to 14 atoms in a chain; germanium, 9; and tin, 2 or 3 only, largely in hydrides (compounds containing hydrogen). Double and triple bonds in catenated arrangements are limited to carbon.

Silicon chains

Catenation, via single or multiple bonds or both, combined with several other factors allows carbon to form more compounds than any other element. These factors are: (1) the stability of certain carbon bonds, in particular of the C—H bond; (2) the existence of carbon in both sp^2 and sp^3 hybridizations; (3) the ability of carbon to form both chain and cyclic compounds (in which the chain of atoms is joined end to end to form a ring) based upon either carbon atoms alone or carbon atoms in combination with those of other nonmetals (*e.g.*, oxygen, sulfur, nitrogen) and either upon single- or multiple-bond arrangements; and (4) the capability of many carbon compounds to exist in isomeric forms (isomers are molecules with identical numbers of the same atoms bonded in different arrangements; such molecules have quite different properties). All but a very few carbon compounds are called organic compounds, and they are discussed in the article CHEMICAL COMPOUNDS.

Atomic size. Reference has been made to some of the physical properties of the carbon group elements (Table 17). Most of the variations in properties from carbon through lead parallel increase in atomic size and are comparable with elements in the boron, nitrogen, oxygen, and fluorine groups. The general trends apparent in Table 17 are roughly those found for the adjacent boron group and nitrogen group elements. The significantly higher melting and boiling points of the carbon group elements reflect their tendency to exist as giant molecules, as opposed to the tendencies of elements in the adjacent families to exist as smaller, discrete molecules.

Special physical properties of carbon

As is true of the lightest element in each group of elements, the physical properties of carbon differ substantially from those of the other members of its family. To a large degree, these differences reflect the substantially higher concentration of the positive charge on the carbon nucleus relative to the size of the carbon atom. That is, the nucleus of carbon holds only six electrons in two shells and, therefore, holds them close; the nucleus of lead, on the other hand, has 82 electrons distributed in six shells. The attraction between the nucleus of lead and its outermost electrons is less than in carbon, because intervening shells in lead shield the outer electrons. Structural differences between diamond and graphite produce profound differences between them in hardness, conductivity, density, heat capacity, and other properties. Inasmuch as graphite is a unique crystalline formation among the elements, its properties should not be compared directly with those of the other elements in the family.

Reactions. Certain of the more common chemical reactions of the carbon group elements are summarized in Table 18. With a given reagent, diamond is generally less reactive than graphite and, thus, requires more rigorous conditions for reaction, such as a higher temperature; the ultimate products, however, are the same. Crystalline silicon is less reactive than finely divided and, possibly, amorphous silicon. Elemental germanium resembles silicon quite closely. Tin and lead behave in general as metals and thus yield at least some ionic products in reactions that are quite different from those of the other elements. Elemental carbon is of particular importance as a high-temperature reducing agent (a reagent that donates electrons) in metallurgical processing for metal oxides, a reaction that frees the metal. Thus to cite only a few of carbon's more important applications, carbon is used directly in the production of elemental phosphorus, arsenic, bismuth, tin, lead, zinc, and cadmium, and indirectly, as carbon monoxide, in the production of iron. Elemental silicon, in the iron–silicon alloy ferrosilicon, is also a strong reducing agent and has been used as such to liberate magnesium from its oxide.

Carbon in metallurgy

Biological and physiological significance. The biological implications of carbon are so extensive that they can be discussed here only very briefly. All biological substances are based upon compounds in which carbon is combined with other elements, the nature of the combination determining the characteristics, function, and relative importance of the substance. To begin with the inorganic compounds of carbon, elemental carbon is nontoxic. Carbon monoxide (CO) is both more readily absorbed and more firmly bound to the hemoglobin of the blood than is oxygen and is thus, even in small concentrations, a dangerous asphyxiant. Carbon dioxide (CO_2) is an asphyxiant of significance only in relatively large concentrations; in small concentration, it stimulates breathing. Hydrogen cyanide (HCN) and its derivatives (cyanogen compounds, cyanides) are all very toxic as protoplasmic poisons through the inhibition of tissue oxidation. Carbon tetrachloride (CCl_4) and other chlorinated hydrocarbons damage the nervous system. Among organic compounds, the most toxic are derivatives that contain the halogen

Table 18: Reactions of the Elements
key: x's—excess; NR—no reaction; g—gas; s—solid; aq.—aqueous; dil.—dilute; conc.—concentrated

reagent	product(s)				
	C	Si	Ge	Sn	Pb
F_2	CF_4	SiF_4	GeF_4	SnF_4	PbF_4
Cl_2	CCl_4	$SiCl_4$ (450° C)	$GeCl_4$	$SnCl_4$	$PbCl_4 \rightarrow PbCl_2$
Br_2	CBr_4	$SiBr_4$	$GeBr_4$	$SnBr_4$	$PbBr_4 \rightarrow PbBr_2$
I_2	CI_4	SiI_4	GeI_4	SnI_4	PbI_2
O_2	CO (x's C) CO_2 (x's O_2)	SiO (x's Si) SiO_2 (x's O_2)	GeO (x's Ge, 850° C) GeO_2 (x's O_2)	SnO_2	PbO (>550° C) Pb_3O_4 (430° C)
S	CS_2 (>1,000° C)	SiS_2	GeS_2 (1,100° C)	SnS_2	PbS
H_2O	CO + H_2 (1,000° C)	$SiO_2 + H_2$			
H_3O^+ (dil. HCl, HBr, H_2SO_4, etc.)	NR	NR	NR	$Sn^{2+} + H_2$	$Pb^{2+} + H_2$
HNO_3 (hot, conc.)	CO_2 (g)	SiO_2 (s)	GeO_2 (s)	SnO_2 (s)	Pb^{2+}
NaOH (hot, aq.)	NR	$SiO_4^{4-} + H_2$	$HGeO_3^- + H_2$	$Sn(OH)_6^{2-} + H_2$	$HPbO_2^- + H_2$
HF (aq.)	NR	$H_2SiF_6 + H_2$	$H_2GeF_6 + H_2$	$SnF_2 + H_2$	$PbF_2 + H_2$
Metal oxide	metal + CO or CO_2	metal + SiO_2	—	—	—
Metals	carbides	silicides	germanides	alloys	alloys

elements (fluorine, chlorine, bromine and iodine), sulfur, selenium, tellurium, nitrogen, phosphorus, arsenic, lead, and mercury. Most organometallic compounds are toxic, while oxygen-containing derivatives of the hydrocarbons are usually less toxic.

Elemental silicon and most silicon-containing compounds appear to be nontoxic. Indeed, human tissue often contains 6 to 90 milligrams of silica (SiO_2) per 100 grams dry weight, and many plants and lower forms of life assimilate silica and use it in their structures. Inhalation of dusts containing alpha SiO_2, however, produces a serious lung disease called silicosis, common among miners, stonecutters, and ceramic workers, unless protective devices are used.

The toxicology of germanium and its compounds is poorly defined.

Elemental tin is apparently nontoxic, and quantities of tin up to 300 parts per million, as dissolved by foods packaged in tin-plated containers and cooking utensils, are not harmful. Organic tin compounds, however, commonly used as biocides and fungicides, are toxic to human beings.

etra-
thyllead

Although elemental lead and difficultly soluble lead compounds are not absorbed by human tissue and are, therefore, quite innocuous, any soluble lead compound is toxic, with toxicity increasing as solubility increases. Symptoms of lead poisoning include abdominal pain and diarrhea followed by constipation, nausea, vomiting, dizziness, headache, and general weakness. Elimination of contact with a lead source is normally sufficient to effect a cure. The elimination of lead from insecticides and paint pigments and the use of respirators and other protective devices in areas of exposure have reduced lead poisoning materially. Low-level atmospheric and water pollution, stemming largely from the use of tetraethyllead, $[Pb(C_2H_5)_4]$ in motor fuels, has been receiving increasing attention.

INDIVIDUAL CARBON GROUP ELEMENTS

Carbon. The word carbon probably derives from the Latin *carbo,* meaning variously "coal," "charcoal," "ember." The term diamond, a corruption of the Greek word *adamas,* "the invincible," aptly describes the permanence of this crystallized form of carbon, just as graphite, the name for the other crystal form of carbon, derived from the Greek verb *graphein,* "to write," reflects its property of leaving a dark mark when rubbed on a surface. Before the discovery in 1779 that graphite when burned in air forms carbon dioxide, graphite was confused with both the metal lead and a superficially similar substance, the mineral molybdenite.

Occurrence and distribution. On a weight basis, carbon is 19th in order of elemental abundance in the crust of the Earth, and there are estimated to be 3.5 times as many carbon atoms as silicon atoms in the universe. Only hydrogen, helium, oxygen, neon, and nitrogen are atomically more abundant in the Cosmos than carbon. Carbon is the cosmic product of the "burning" of helium in which three helium nuclei, atomic number 4, fuse to produce a carbon nucleus, atomic number 12. In the crust of the Earth, elemental carbon is a minor component: carbon compounds (*i.e.,* carbonates of magnesium and calcium) form common minerals (*e.g.,* magnesite, dolomite, marble, or limestone). Coral and the shells of oysters and clams are primarily calcium carbonate. Carbon is widely distributed as coal and in the organic compounds that constitute petroleum, natural gas, and all plant and animal tissue. A natural sequence of chemical reactions called the carbon cycle—involving conversion of atmospheric carbon dioxide to carbohydrates by photosynthesis in plants, the consumption of these carbohydrates by animals and oxidation of them through metabolism to produce carbon dioxide and other products, and the return of carbon dioxide to the atmosphere—is one of the most important of all biological processes.

Carbon
in all
living
tissue

Production of elemental carbon. Elemental carbon is best considered in terms of its several crystalline forms—diamond, graphite, and "amorphous carbon"—since the widely different properties of these three substances require different approaches. (When an element exists in more than one crystal form, each is called an allotrope.)

Until 1955, all diamonds were obtained from natural deposits, most significant in southern Africa but occurring also in Brazil, Venezuela, British Guiana (now Republic of Guyana), and Siberia. The single known source in the United States, in Arkansas, has no commercial importance, nor is India, once historically a source of fine diamonds, a significant present-day supplier. The primary source of diamonds is a soft, bluish-coloured peridotic rock called kimberlite (after the famous deposit at Kimberley, South Africa), found in volcanic structures called pipes; but many diamonds occur in alluvial deposits presumably resulting from the weathering of primary sources. Isolated finds around the world in regions where no sources are indicated have not been uncommon. Natural deposits are worked by crushing, by gravity and flotation separations, and by removal of diamonds by their adherence to a layer of grease on a suitable table. The following products result: (1) diamond proper—distorted cubic-crystalline, gem-quality stones varying from colourless to red, pink, blue, green, and yellow; (2) bort—minute, dark crystals of abrasive but not gem quality; (3) ballas—randomly oriented crystals of abrasive quality; (4) macles—triangular, pillow-shaped crystals that are industrially useful; and (5) carbonado—mixed diamond–graphite crystallites containing other impurities.

Artificial
diamonds

The successful laboratory conversion of graphite to diamond was made in 1955. The procedure involved the simultaneous use of extremely high pressure and temperature with iron as a solvent or catalyst. Subsequently, chromium, manganese, cobalt, nickel, and tantalum were substituted for iron. Synthetic diamonds are now manufactured in several countries and are being used increasingly in place of natural materials as industrial abrasives.

Graphite occurs naturally in many areas, the deposits of major importance being in the Republic of Korea, Austria, China, Mexico, Madagascar, Germany, Sri Lanka, and the Soviet Union. Both surface- and deep-mining techniques are used, followed by flotation, but the major portion of commercial graphite is produced by heating petroleum coke in an electric furnace. A better crystallized form, known as pyrolytic graphite, is obtained from the decomposition of low-molecular-weight hydrocarbons by heat. Graphite fibres of considerable tensile strength are obtained by carbonizing natural and synthetic organic fibres.

Amorphous carbon, which is solid but not clearly crystallized, probably consists of microcrystals of graphite; it is common in commerce as coke; lampblack, or carbon black; and charcoal. These products are obtained by heating coal (to give coke), natural gas (to give blacks), or carbonaceous material of vegetable or animal origin, such as wood or bone (to give charcoal), at elevated temperatures in the presence of insufficient oxygen to allow combustion. The volatile by-products are recovered and used separately.

Structure of carbon allotropes. The crystal structure of diamond is an infinite three-dimensional array of carbon atoms, each of which forms a structure in which each of the bonds makes equal angles with its neighbours. If the ends of the bonds are connected, the structure is that of a tetrahedron, a three-sided pyramid of four faces (including the base). Every carbon atom is covalently bonded at the four corners of the tetrahedron to four other carbon atoms. The distance between carbon atoms along the bond is 1.54×10^{-8} centimetre, and this is called the single-bond length. The space lattice of the diamond can be visualized as carbon atoms in puckered hexagonal (six-sided) rings that lie roughly in one plane, the natural cleavage plane of the crystal; and these sheets of hexagonal, puckered rings are stacked in such a way that the atoms in every fourth layer lie in the same position as those in the first layer. The layer arrangement sequence is thus ABCABCA.... Such a crystal structure can be destroyed only by the rupture of many strong bonds: thus the extreme hardness, high sublimation temperature, the presumed extremely high melting point (extrapolated from known behaviour), and reduced chemical reactivity and insulating properties are all reasonable consequences of the crystal structure. Because of both the sense and the direction of the tetrahedral axis, four spatial orientations of carbon atoms exist, leading to two tetrahedral and two octahedral (eight-faced) forms of diamond.

Crystal structure of graphite

The crystal structure of graphite amounts to a parallel stacking of layers of carbon atoms. Within each layer the carbon atoms lie in fused hexagonal rings that extend infinitely in two dimensions. The stacking pattern of the layers is ABABA . . . ; that is, each layer separates two identically oriented layers. Within each layer the carbon–carbon bond distance is 1.42×10^{-8} centimetre, which is intermediate between the single bond and double (1.33×10^{-8} centimetre) bond distances. All carbon–carbon bonds within a layer are the same (an observation that is interpreted in terms of complete π-bonding). The interlayer distance (3.37×10^{-8} centimetre) is sufficiently large to preclude localized bonding between the layers; the bonding between layers is probably by van der Waals interaction (*i.e.,* the result of attraction between electrons of one carbon atom and the nuclei of neighbouring atoms). Ready cleavage, as compared with diamond, and electrical conductivity are consequences of the crystal structure of graphite. Other related properties are softness and lubricity (smoothness, slipperiness). A less common form of graphite, which occurs in nature, is based upon an ABCABCA . . . stacking, in which every fourth layer is the same. The amorphous varieties of carbon are based upon microcrystalline forms of graphite.

The greater degree of compactness in the diamond structure as compared with graphite suggests that by the application of sufficient pressure on graphite it should be converted to diamond. At room temperature and atmospheric pressure, diamond is actually less stable than graphite. The rate of conversion of diamond to graphite is so slow, however, that a diamond persists in its crystal form indefinitely. As temperature rises, the rate of conversion to graphite increases substantially, and at high temperatures it becomes (thermodynamically) favourable if the pressure is sufficiently high. At the same time, however, the rate of conversion decreases as the (thermodynamic) favourability increases. Thus, graphite does not yield diamond when heated under high pressure, and it appears that direct deformation of the graphite structure to the diamond structure in the solid state is not feasible. The occurrence of diamonds in iron–magnesium silicates in the volcanic structures called pipes and in iron–nickel and iron sulfide phases in meteorites suggests that they were formed by dissolution of carbon in those compounds and subsequent crystallization from them in the molten state at temperatures and pressures favourable to diamond stability. The successful synthesis of diamond is based upon this principle.

Diamonds in meteorites

The temperature and pressure relationships of any substance can be plotted to show how, for example, the boiling point changes as pressure is changed; such graphs are called phase diagrams (see MATTER). They reveal conditions under which rearrangements in the atomic or molecular structure of a substance take place.

The crystal structure of graphite is of a kind that permits the formation of many compounds, called lamellar or intercalation compounds, by penetration of molecules or ions. Graphitic oxide and graphitic fluoride are nonconducting lamellar substances not obtained in true molecular forms that can be reproduced, but their formulas do approximate, respectively, the compositions of carbon dioxide and carbon monofluoride.

A type of chemical reaction in which one substance (an oxidizing agent) accepts electrons from another substance (a reducing agent) and is thereby reduced (while the reducing agent is oxidized) is frequently observed with carbon and its compounds. Although carbon is usually a reducing agent, under acidic conditions elemental carbon is a moderately strong oxidizing agent. The large energy of the carbon–carbon bond makes activation energy requirements for the reaction so high that direct reduction of carbon—*e.g.,* to methane (formula CH_4)—is impractical. Reduction of carbon monoxide to elemental carbon and oxidation of carbon monoxide to carbon dioxide are both feasible but impractical in solution. Under alkaline conditions, only the oxidation of formate ion (HCO_2^-) to carbonate ion (CO_3^{2-}) is a reasonable process.

Nuclear properties. The notation used for the nucleus of atoms places the atomic mass as a presuperscript to the symbol of the element and the atomic number as a presubscript; thus, the isotope carbon-12 is symbolized ${}^{12}_{6}C$. Of the stable nuclides (Table 17), the isotope carbon-13 is of particular interest in that its nuclear spin imparts response in a device called a nuclear magnetic resonance spectrometer, which is useful when investigating the molecular structures of covalently bonded compounds containing carbon. This isotope is also useful as a label in compounds that are to be analyzed by mass spectrometry, another device that is used extensively to identify atoms and molecules. Of the unstable nuclides, only carbon-14 is of sufficiently long half-life to be important. It is formed by the interaction of neutrons, produced by cosmic radiation, with nitrogen (N) in the atmosphere in a reaction that may be written as follows (neutron is symbolized as ${}^{1}_{0}n$, the nitrogen atom as ${}^{14}_{7}N$, and a hydrogen nucleus, or proton, as ${}^{1}_{1}H$):

$${}^{14}_{7}N + {}^{1}_{0}n \longrightarrow {}^{14}_{6}C + {}^{1}_{1}H$$

The carbon-14 atoms from this reaction are converted to carbon dioxide by reaction with atmospheric oxygen and mixed and uniformly distributed with the carbon dioxide containing stable carbon-12. Living organisms use atmospheric carbon dioxide, whether with stable or radioactive carbon, through processes of photosynthesis and respiration, and thus their systems contain the constant ratio of carbon-12 to carbon-14 that exists in the atmosphere.

Death of an organism terminates this equilibration process; no fresh carbon dioxide is added to the dead substance. The carbon-14 present in the dead substance decays in accordance with its 5,730-year (± 40 years) half-life, while the carbon-12 remains what it was at death. Measurement of the carbon-14 activity at a given time thus allows calculation of the time elapsed after the death of the organism. Measurement of the carbon-14 activity in a cypress beam in the tomb of the Egyptian Pharaoh Snefru, for example, established the date of the tomb as *c.* 2600 BC. Many items of archaeological significance have been dated similarly.

The nuclides carbon-12 and carbon-13 are of importance in the carbon cycle of energy creation in certain stars. The cycle can be summarized in terms of nuclear equations, the separate steps being:

$${}^{12}_{6}\mathrm{C} + {}^{1}_{1}\mathrm{H} \longrightarrow {}^{13}_{7}\mathrm{N} + \text{energy}$$

$${}^{13}_{7}\mathrm{N} \longrightarrow {}^{13}_{6}\mathrm{C} + {}^{0}_{+1}e$$

$${}^{13}_{6}\mathrm{C} + {}^{1}_{1}\mathrm{H} \longrightarrow {}^{14}_{7}\mathrm{N} + \text{energy}$$

$${}^{14}_{7}\mathrm{N} + {}^{1}_{1}\mathrm{H} \longrightarrow {}^{15}_{8}\mathrm{O} + \text{energy}$$

$${}^{15}_{8}\mathrm{O} \longrightarrow {}^{15}_{7}\mathrm{N} + {}^{0}_{+1}e$$

$${}^{15}_{7}\mathrm{N} + {}^{1}_{1}\mathrm{H} \longrightarrow {}^{12}_{6}\mathrm{C} + {}^{4}_{2}\mathrm{He}.$$

Summation of the equations allows the fusion process to be written as a reaction among four atoms of hydrogen to yield one atom of helium (He), two positrons (${}^{0}_{+1}e$), and energy:

$$4\,{}^{1}_{1}\mathrm{H} \longrightarrow {}^{4}_{2}\mathrm{He} + 2\,{}^{0}_{+1}e + \text{energy};$$

this equation does not show that the process uses up and regenerates the carbon-12. In a sense, carbon acts as a catalyst for this mode of converting mass to energy.

Biological implications. The biological implications of the element and its simple compounds have been discussed earlier; the most significant implications are associated with organic compounds, which are discussed in such articles as CHEMICAL COMPOUNDS; BIOCHEMICAL COMPONENTS OF ORGANISMS.

Analysis. Carbon, either elemental or combined, is usually determined quantitatively by conversion to carbon dioxide gas, which can then be absorbed by other chemicals to give either a weighable product or a solution with acidic properties that can be titrated.

Silicon. The name silicon derives from the Latin *silex* or *silicis,* meaning "flint" or "hard stone."

Occurrence and distribution. On a weight basis, the abundance of silicon in the crust of the Earth is exceeded only by oxygen. Estimates of the cosmic abundance of other elements often are cited in terms of the number of their atoms per 10^6 atoms of silicon. Only hydrogen, helium, oxygen, neon, nitrogen, and carbon exceed silicon in cosmic abundance. Silicon is believed to be a cosmic product of alpha-particle absorption, at a temperature of about 10^9 K, by the nuclei of carbon-12, oxygen-16, and neon-20. The energy binding the particles that form the nucleus of silicon is about 8.4 million electron volts (MeV) per nucleon (proton or neutron). Compared with the maximum of about 8.7 million electron volts for the nucleus of iron, almost twice as massive as that of silicon, this figure indicates the relative stability of the silicon nucleus. The pure element silicon is too reactive to be found in nature. In compounds, the oxidized form, as silicon dioxide and particularly as silicates, is common in the Earth's crust

Silicon minerals

and is an important component of the Earth's mantle. Silicon dioxide occurs both in crystalline minerals (*e.g.,* quartz, cristobalite, tridymite) and amorphous or seemingly amorphous minerals (*e.g.,* agate, opal, chalcedony) in all land areas. The natural silicates are characterized by their abundance, wide distribution, and structural and compositional complexities. Most of the elements of the following groups in the periodic table are found in silicate minerals: groups Ia, IIa, IIIa, IIIb, IVb, Vb, VIb, VIIa. These elements are said to be lithophilic, or stone-loving. Important silicate minerals include the clays, feldspar, olivine, pyroxene, amphiboles, micas, and zeolites.

Production of elemental silicon. Elemental silicon is produced commercially by the reduction of silica (SiO_2) with coke in an electric furnace, and the impure product is then refined. Almost pure silicon is obtained by the reduction of silicon tetrachloride or trichlorosilane. For use in electronic devices, single crystals are grown by slowly withdrawing seed crystals from molten silicon.

Properties of the element. The more important physical and chemical properties of silicon have been summarized and related to the ground-state electronic configuration of the silicon atom and the position of silicon in the periodic table. Because silicon forms chains similar to those formed by carbon, silicon has been studied as a possible base element for silicon organisms. Another property resulting from the electronic structure of silicon is that it functions as an intrinsic semiconductor (see MATTER). Addition of an element such as boron, an atom of which can be substituted for a silicon atom in the crystal structure but which provides one less valence electron (boron is an acceptor atom) than silicon, allows silicon atoms to lose electrons to it. The positive holes created by the shift in electrons allow extrinsic semiconduction of a type referred to as positive (*p*). Addition of an element such as arsenic, an atom of which can also be substituted for a silicon atom in the crystal but which provides an extra valence electron (arsenic is a donor atom), releases its electron within the lattice. These electrons allow semiconduction of the negative (*n*) type. If *p*-silicon and *n*-silicon wafers are joined, in a manner called the *p–n* junction, and placed in sunlight,

The solar cell

the absorbed energy causes electrons to move across the junction and an electric current to flow in an external circuit connecting the two wafers. Such a solar cell is a source of energy for space devices.

Unlike carbon, the only crystalline form in which silicon exists is an octahedral form based upon atoms in the diamond-type arrangement. The amorphous forms of silicon contain micro crystals of this type. The reduced bond energy in crystalline silicon renders the element lower melting, softer, and chemically more reactive than diamond.

Only the 0 and +4 oxidation states of silicon are stable in aqueous systems.

As is true with carbon, the bonds in elemental silicon are strong enough to require large energies to activate, or promote, reaction in an acidic medium, and the oxidation–reduction reactions do not appear to be reversible at ordinary temperatures.

Germanium. The name germanium derives from the Latin word Germania (Germany) and was given to the element by its German discoverer, Clemens Winkler, in 1886.

Occurrence and distribution. On a weight basis, germanium is a scarce but not an extremely rare element in the crust of the Earth, equalling in abundance beryllium, molybdenum, and cesium and exceeding the elements arsenic, cadmium, antimony, and mercury. In the Cosmos, the atomic abundance of germanium is 50.5 (based upon $Si = 1 \times 10^6$), a value roughly equal to those for krypton and zirconium and only slightly less than that for selenium. The cosmic abundance is much less than those of a number of the heavier elements; *e.g.,* bromine, strontium, tin, barium, mercury, and lead. All of the elements of lower nuclear charge than germanium, except beryllium, boron, scandium, and gallium, are cosmically more abundant than germanium. Cosmically, germanium is believed to be one of the many elements formed by neutron absorption after the initial processes of hydrogen and helium burning and alpha-particle absorption.

Natural sources of germanium

Germanium is widely distributed in nature but is too reactive to occur free. Primary minerals include argyrodite, germanite, renierite, and canfieldite, all of them rare; only germanite and renierite have been used as commercial sources for the element. Trace quantities of germanium are found in certain zinc blendes, in sulfidic ores of copper and arsenic, and in coals, the latter possibly a consequence of the concentration of the element by plants of the Carboniferous Period in geologic history. Certain present-day plants are known to concentrate germanium. Both zinc-process concentrates and ash and flue dusts from coal-burning installations provide commercial sources of germanium.

Production of germanium. Pure germanium is obtained from these sources in a complicated process that finally yields billets or blocks, which may be purified by further refining to the quality required for the manufacture of semiconductors. Single crystals of germanium are grown in an atmosphere of nitrogen or helium from the molten material. These are then transformed into semiconductors by being doped (infused) with electron donor or acceptor atoms, either by incorporating the impurities in the melt during growth of the crystal or by diffusing the impurities into the crystal after it has been formed.

Properties of the element. The physical and chemical properties of germanium have been summarized earlier. Like silicon, germanium crystallizes in the diamond type of structure. The trend in properties noted with silicon and related to reduced bond energy continues with germanium. The electrical and semiconducting characteristics of germanium are comparable to those of silicon.

Germanium(II) compounds are well characterized as solids, and in general they are readily oxidized. Elemental germanium can be electrodeposited from many solutions and melts of its compounds. It is of interest that as little as one milligram of dissolved germanium per litre seriously interferes with the electrodeposition of zinc.

Tin. The symbol Sn for tin is an abbreviation of the Latin word for tin, *stannum.*

Occurrence and distribution. On a weight basis, tin is a scarce but not rare element in the crust of the Earth, its abundance being of the same order of magnitude as such technically useful elements as cobalt, nickel, copper, cerium, and lead, and it is essentially equal to the abundance of nitrogen. In the Cosmos, there are 1.33 atoms of tin per 1×10^6 atoms of silicon, an abundance roughly equal to that of niobium, ruthenium, neodymium, or platinum. Cosmically, tin is a product of neutron absorption. Its richness in stable isotopes is noteworthy.

Tin in Phoenician times

The only mineral of commercial significance is cassiterite (SnO_2). No high-grade deposits are known. The major sources are alluvial deposits, averaging about 0.01 percent tin. Lode deposits, containing up to 4 percent, are found in Bolivia and Cornwall. The latter were worked at least as early as Phoenician times but are no longer of major consequence. Some 90 percent of world production comes from Malaysia, Thailand, Indonesia, Bolivia, Congo (Brazzaville), Nigeria, and China. The United States has no significant deposits. (For commercial production see INDUSTRIES, EXTRACTION AND PROCESSING: *Production of other important industrial metals.*)

Properties of the element. The major physical and chemical properties of tin have been summarized earlier. The relationships among the allotropic modifications of tin can be represented as transformations from one crystal type to another at specific temperatures:

$$\underset{\text{(cubic-diamond crystals)}}{\text{gray or alpha } (\alpha) \text{ tin}} \xrightleftharpoons{13.2^\circ \text{ C } (55.8^\circ \text{ F})} \underset{\text{(tetragonal crystals)}}{\text{white or beta } (\beta) \text{ tin}}$$

$$\xrightleftharpoons{161^\circ \text{ C } (322^\circ \text{ F})} \underset{\text{(rhombic crystals)}}{\text{brittle tin}} .$$

(The double arrows signify that the transformation occurs in both directions, as tin is heated or as it is cooled.) Gray tin is easily crumbled, whereas white tin is malleable and ductile. The spontaneous conversion of the bright metal to a gray powder at low temperatures ("tin pest") seriously hampers the use of the metal in very cold regions. This change is rapid only below −50° C, unless catalyzed by gray tin or tin(IV). As a consequence of internal friction, white tin emits a characteristic sound ("tin cry") when bent. Elemental tin is readily oxidized to the dipositive ion in acidic solution, but this tin(II) ion is converted to tin(IV) ion by many mild oxidizing agents, including elemental oxygen. Oxidation under alkaline conditions normally gives the tetrapositive (Sn^{4+}) state. In an alkaline medium, dipositive tin (Sn^{2+}) disproportionates readily to tetrapositive tin and the free element.

Lead. The symbol Pb for lead is an abbreviation of the Latin word for lead, *plumbum.*

Occurrence and distribution. On a weight basis, lead has nearly the same abundance in the Earth's crust as tin. Cosmically, there is 0.47 lead atom per 10^6 silicon atoms. The cosmic abundance is comparable with those of cesium, praseodymium, hafnium, and tungsten, each of which is regarded as a reasonably scarce element. Lead is formed both by neutron-absorption processes and the decay of radionuclides of heavier elements. Stable lead nuclides are the end products of radioactive decay in the three natural decay series: uranium (decays to lead-206), thorium (decays to lead-208), and actinium (decays to lead-207).

Although lead is not abundant, natural concentration processes have resulted in substantial deposits of commercial significance, particularly in the United States, but also in Canada, Africa, Australia, Spain, Germany, and South America. Significant deposits are found in the U.S. in the western states and the Mississippi Valley. Elemental lead is only rarely found in nature. The major minerals are galena (PbS), anglesite ($PbSO_4$), and cerussite ($PbCO_3$). (For commercial production see INDUSTRIES, EXTRACTION AND PROCESSING: *Lead production.*)

Useful physical properties of lead

Properties of the element. The major physical and chemical properties of lead and their relationships to electronic configuration and position in the periodic table have been summarized earlier. Only a single crystalline modification, with a close-packed metallic lattice, is known. Properties that are responsible for the many uses of elemental lead include its ductility, ease of welding, low melting point, high density, and ability to absorb gamma radiation and X-radiation. Molten lead is an excellent solvent and collector for elemental silver and gold. The structural applications of lead are limited by its low tensile and fatigue strengths and its tendency to flow even when only lightly loaded. Elemental lead can be oxidized to lead(II) ion by hydrogen ion, but the insolubility of most lead(II) salts makes lead resistant to attack by many acids. Oxidation under alkaline conditions is easier to effect and is favoured by the formation of the soluble lead(II) species. Lead(IV) oxide is among the stronger oxidizing agents in acidic solution, but it is comparatively weak in alkaline solution. The ease of oxidation of lead is enhanced by complex formation. The electrodeposition of lead is best effected from aqueous solutions containing lead(II) hexafluorosilicate(IV) and hexafluorosilicic acid.

Nuclear properties. Of the radionuclides of lead, the following appear as members of the three natural decay series: (1) thorium series: lead-212; (2) uranium series: lead-214 and lead-210; (3) actinium series: lead-211. The atomic weight of natural lead varies from source to source, depending on its origin by heavier element decay.

(T.Mo./Ed.)

Nitrogen group elements

In the formal classification of the chemical elements known as the periodic table, the nitrogen group elements, consisting of nitrogen (N), phosphorus (P), arsenic (As), antimony (Sb), and bismuth (Bi), form a vertical file commonly designated group Va (see Figure 10). These elements share certain general similarities in chemical behaviour, though they are clearly differentiated from one another chemically, and these similarities reflect common features of the electronic structures of their atoms.

Probably no other group of the elements is more familiar to the layman than this group. Although the five elements together make up less than 0.2 percent by weight of the Earth's crust, they assume an importance far out of proportion to their abundance. This is especially true of the elements nitrogen and phosphorus, which comprise 2.4 and 0.9 percent, respectively, of the total weight of the human body.

The nitrogen elements have, perhaps, the widest range in physical state of any group in the periodic table. Nitrogen, for example, is a gas that liquefies at about −200° C, whereas bismuth is a solid melting at 271° C and boiling at about 1,560° C. Chemically, too, the range in properties is wide, nitrogen and phosphorus being typical nonmetals; arsenic and antimony, metalloids; and bismuth, a metal. Even in appearance these elements exhibit great variety. Nitrogen is colourless both as a gas and as a liquid. Phosphorus exists in a variety of physical modifications, or allotropic forms, including the familiar white, highly

Physical appearance

reactive form that must be stored under water to prevent it from igniting in the air; a much less reactive red or violet form; and a black modification that, although least known, appears to be the most stable of all. Arsenic exists mainly as a dull gray metallic solid, but a more reactive yellow, solid form is also known, and there are indications that other forms exist under certain conditions. Antimony is a silver, metallic appearing, but somewhat brittle solid; and bismuth is a silver-white metal with a trace of pink in its lustre.

Together with carbon, hydrogen, oxygen, and sulfur, the first two members of this group, nitrogen and phosphorus, are the principal chemical elements incorporated into living systems. Nitrogen and phosphorus are readily removed from the soil by plant growth, and therefore they are immensely important components of plant foods. Such designations as "5–10–5" on commercial fertilizers represent the respective weight percentage composition of the material in terms of nitrogen, phosphoric oxide, and potassium oxide (potassium being the third principal element needed for healthy plant growth). Nitrogen in fertilizers may be in the form of sodium or potassium nitrates, ammonia, ammonium salts, or various organic combinations. Phosphorus is supplied chiefly as inorganic phosphate.

These same elements, nitrogen and phosphorus, can also be used in ways less helpful to man. The explosives in conventional warfare are heavily dependent on their content of nitrogen compounds, and the deadly nerve gases are composed of organic compounds of phosphorus.

On the other hand, arsenic, which is notorious for its toxicity, is most useful in agriculture, where its compounds are an aid in controlling harmful insect pests. Antimony and bismuth are used chiefly in metal alloys, because they impart unique and desirable properties to these alloys.

HISTORY

Early theories of air

Nitrogen. About four-fifths of the Earth's atmosphere is nitrogen, which was isolated and recognized as a specific substance during early investigations of the air. Carl Wilhelm Scheele, a Swedish druggist, showed in 1772 that air is a mixture of two gases, one of which he called "fire air," because it supported combustion, and the other "foul air," because it was left after the "fire air" had been used up. The "fire air" was, of course, oxygen, and the "foul air" nitrogen. At about the same time nitrogen also was recognized by a Scottish botanist, Daniel Rutherford, and by the controversial British clergyman, Joseph Priestley, who, with Scheele, is given credit for the discovery of oxygen. Later work showed the new gas to be a constituent of nitre, a common name for potassium nitrate (KNO_3); and, accordingly, it was named nitrogen by the French chemist Jean-Antoine-Claude Chaptal in 1790. Nitrogen also was considered a chemical element by Antoine-Laurent Lavoisier, whose explanation of the role of oxygen in combustion eventually overthrew the phlogiston theory, an erroneous view of combustion that became popular in the early 18th century. The inability of nitrogen to support life led Lavoisier to name it *azote,* still the French equivalent of nitrogen.

Phosphorus. Arabian alchemists of the 12th century may have isolated elemental phosphorus by accident, but the records are unclear. Phosphorus appears to have been discovered in 1669 by Henning Brand, a German merchant whose hobby was alchemy. Brand allowed 50 buckets of urine to stand until they putrified and "bred worms." He then boiled the urine down to a paste and heated it with sand, thereby distilling elemental phosphorus from the mixture. Brand reported his discovery in a letter to Gottfried Wilhelm Leibniz, and, thereafter, demonstrations of this element and its ability to glow in the dark, or "phosphoresce," excited public interest. Phosphorus, however, remained a chemical curiosity until about a century later when it proved to be a component of bones. Digestion of bones with nitric or sulfuric acid formed phosphoric acid, from which phosphorus could be distilled by heating with charcoal. In the late 1800's, James Burgess Readman of Edinburgh developed an electric furnace method for producing the element from phosphate rock, which is essentially the method employed today.

Arsenic. Arsenic was known in the form of certain of its compounds long before it was clearly recognized as a chemical element. In the 4th century BC Aristotle wrote of a substance called *sandarachē,* now believed to have been the mineral realgar, a sulfide of arsenic. Then, in the 1st century AD, the writers Pliny the Elder and Pedanius Dioscorides both described *auripigmentum,* a substance now thought to have been the dyestuff orpiment, As_2S_3. By the 11th century AD three species of "arsenic" were recognized: white (As_4O_6), yellow (As_2S_3), and red (As_4S_4). The element itself possibly was first observed in the 13th century by Albertus Magnus, who noted the appearance of a metallike substance when *arsenicum,* another name for As_2S_3, was heated with soap. It is not certain, however, that this natural scientist and scholar actually observed the free element. The first clearly authentic report of the free substance was made in 1649 by Johann Schroeder, a German pharmacist, who prepared arsenic by heating its oxide with charcoal. Later, Nicolas Lêmery, a French physician and chemist, observed the formation of arsenic when heating a mixture of the oxide, soap, and potash. By the 18th century, arsenic was well known as a unique semimetal.

Antimony. The ancients were familiar with antimony both as a metal and in its sulfide form. Fragments of a Chaldean vase made of antimony have been estimated to date from about 4000 BC. The Old Testament tells of Queen Jezebel using the naturally occurring sulfide of antimony to beautify her eyes. Pliny, during the 1st century AD, wrote of seven different medicinal remedies using *stibium* or antimony sulfide. Early writings of Dioscorides, dating from about the same time, mention metallic antimony. Records of the 15th century show the use of the substance in alloys for type, bells, and mirrors. In 1615 Andreas Libavius, a German physician, described the preparation of metallic antimony by the direct reduction of the sulfide with iron; and a later chemistry textbook by Lêmery, published in 1675, also describes methods of preparation of the element. In the same century, a book summarizing available knowledge of antimony and its compounds was purportedly written by a Basil Valentine, allegedly a Benedictine monk of the 15th century, whose name appears on chemical writings over a span of two centuries. The name antimony appears to be derived from the Latin *antimonium,* in a translation of a work by the alchemist Geber, but its real origin is uncertain.

Bismuth. Bismuth evidently was known in very early times, since it occurs in the native state as well as in compounds. For a long period, however, it was not clearly recognized as a separate metal, being confused with such metals as lead, antimony, and tin. Miners during the Middle Ages apparently believed bismuth to be a stage in the development of silver from baser metals and were dismayed when they uncovered a vein of the metal thinking they had interrupted the process. In the 15th-century writings of Basil Valentine this element is referred to as *wismut.* The name may have been derived from the German words *wis mat,* meaning white mass. In any case it was latinized to *bisemutum* by the mineralogist Georgius Agricola, who recognized its distinctive qualities and described how to obtain it from its ores. Bismuth was accepted as a specific metal by the middle of the 18th century, and works on its chemistry were published in 1739 by the German chemist Johann Heinrich Pott and in 1753 by the Frenchman Claude-François Geoffroy.

COMPARATIVE CHEMISTRY

Electronic configurations. *Similarities in orbital arrangement.* In the periodic table, each of the nitrogen group elements occupies the fifth position among the main group elements of its period, a position designated Va. In terms of the electronic configuration of its atoms, each nitrogen group element possesses an outermost shell of five electrons. In each case, these occupy an outer *s* orbital completely (with two electrons) and contribute one electron to each of the three outer *p* orbitals (the orbitals being electron regions within the atom and the letter designations, *s, p, d,* and *f,* being used to designate different classes of orbital). The arrangement of outer electrons in the atoms of the nitrogen elements thus provides three half-filled outer orbitals that, by interaction with half-filled orbitals of the atoms of other elements, can form three covalent bonds. The other atoms may attract the shared electrons either more or less strongly than do the nitrogen group atoms; therefore the latter may acquire either positive or negative charges and exist in oxidation states of +3 or −3 in their compounds. In this respect, the nitrogen elements are alike.

Lone pairs

Another similarity among the nitrogen elements is the existence of an unshared, or lone, pair of electrons, which remains after the three covalent bonds, or their equivalent, have been formed. This lone pair permits the molecule to act as an electron pair donor in the formation of molecular addition compounds and complexes. The availability of the lone pair depends upon various factors, such as the relative size of the atom, its partial charge in the molecule, the spatial characteristics of other groups in the molecule, and the as-yet poorly understood phenomenon called the "inert pair effect." This effect consists of a tendency for the paired *s* electrons in the outermost shell of the heavier atoms of a major group to remain chemically unreactive. Because of it, the electron pair-donating ability of the nitrogen group elements is not uniform throughout the group; it is probably greatest with nitrogen, less with the intermediate elements, and nonexistent with bismuth.

Variations in bonding capacity. Significant differences in electronic configurations also occur among the elements of the nitrogen group with respect both to the underlying shell and to the outer *d* orbitals. Since the latter first appear with the third period of the table, they are present in all elements of the group but nitrogen. The possibility of utilizing these outer *d* orbitals for bonding thus exists for phosphorus, arsenic, antimony, and bismuth, but not for nitrogen.

Utilization of *d* orbitals

There are three principal ways in which the outer *d* orbitals can be used to increase the number of bonds or expand the valence octet. One is by providing a space to which one of the *s* electrons can be promoted. This creates two additional half-filled orbitals (one *d* and one *s* orbital), and it therefore generates the capacity to form two additional covalent bonds. This is exemplified by the production of phosphorus pentafluoride, PF_5, by further fluorination of the trifluoride, PF_3. Such promotion appears to be greatly assisted by the increase in outer *d*-orbital stability that results from the withdrawal of part of the screening electron and the attendant increase of the effective nuclear charge of the central atom. In PF_5, for example, the fluorine atoms, being much more electronegative than the phosphorus atom, draw away a portion of the phosphorus electrons, leaving the outer *d* orbitals more exposed to the phosphorus nucleus and therefore more stable.

A second way in which the outer *d* orbitals can become involved in the bonding is by their becoming sufficiently stable to attract a lone pair of electrons from a donor. For example, PF_5 can serve as an electron pair acceptor through an outer *d* orbital to coordinate a fluoride ion donor and form the complex ion PF_6^-.

A third way of involving *d* orbitals in bonding is for them to become partially occupied in accommodating lone-pair electrons from another atom, which is already attached by a single bond, thereby strengthening the bond. The phosphorus oxyhalides, of general formula POX_3, appear to be examples of this; their phosphorus–oxygen bonds are observed to be shorter and stronger than expected for ordinary single bonds.

Alternation of properties

The +5 oxidation state. It thus is possible for an atom of phosphorus, arsenic, antimony, or bismuth to expand its valence octet to form five covalent bonds and one additional coordinate covalent bond. This is not possible for nitrogen, which exhibits a maximum coordination number of four: three single covalent bonds and a coordinate covalent bond with nitrogen acting as donor (through its lone pair). Nevertheless, the +5 oxidation state is formally applicable to nitrogen, so that all five elements can be found in this state. When compounds in the +5 oxidation state are studied, however, it is observed that their properties do not exhibit a uniform trend within the group. Rather, a certain degree of alternation is observed, the +5 states of nitrogen, arsenic, and bismuth appearing less stable and more strongly oxidizing than the corresponding states of phosphorus and antimony. In part this alternation may find explanation in the electronic differences among the atoms with respect to their underlying shells. The number of electrons in the shell just below the outermost level, is two for nitrogen, eight for phosphorus, and 18 for arsenic, antimony, and bismuth.

Increasing the nuclear charge by 18 from phosphorus to arsenic may be accompanied by incomplete shielding of this extra charge by the ten 3*d* electrons also added. This would imply smaller size and a greater electronegativity for arsenic than for phosphorus and thus a greater similarity between the phosphorus and antimony atoms. This subject, however, is still controversial, and the widely used scale of electronegativities devised by Linus Pauling fails to make this distinction.

An interesting anomaly is presented by the fact that nitrogen as a free element is in the form of gaseous diatomic molecules, while the elements immediately preceding it in its period of the table are solids, as are the other elements in its group. In surveying the elements of the second period, the most obvious difference in atomic structure found on reaching nitrogen is the appearance for the first time in compounds of the element of a lone pair of electrons not used in bonding with other atoms. Calculations suggest that the presence of this lone pair of electrons is associated with a considerable weakening of nitrogen to nitrogen single bonds in compounds where these bonds occur. In the diatomic nitrogen molecule, however, the bonding is of a different variety—triple bonds being found between the atoms. It is thought that the triple bond is unaffected (unweakened) by the lone pairs of electrons on the nitrogen atoms, and this is assumed to be the reason why nitrogen "prefers" to exist as triply bonded gaseous diatomic molecules rather than as a condensed singly bonded solid polymer.

Bond weakening by lone pairs

The same effect might be expected to be operable with the other elements of the nitrogen group, all of which also contain lone electron pairs in their outermost shells. Further calculations disclose, however, that the bond-weakening effect of the lone pair is far less pronounced with these elements than it is with nitrogen. As a result, with these elements, single bonds are favoured over multiple bonds, and the diatomic state of the molecules is not the preferred form.

Relative electronegativities. It might also be expected that the weakening effect of the lone pair would be observed in compounds of the nitrogen group elements. The picture is more complicated here because the bonds under discussion are formed between different types of atoms. Since different elements differ in electronegativity, bonds between the atoms of different elements are inevitably polar. For purposes of discussion it can be assumed that polar bonds consist of blends of nonpolar covalent bonds and completely polar, ionic bonds. It can then be shown that a relatively small amount of ionic character will contribute a disproportionate share to the overall bond strength. Since the weakening effect of the lone pair is felt only on the covalent portion of the polar bond, rather than on the ionic portion, the less polar bonds will exhibit the greater lone-pair weakening effects.

Comparison of nitrogen group elements. These considerations become important in comparing the chemical behaviour of the nitrogen group elements. The electronegativity of nitrogen itself, although lower than that of oxygen, is substantially higher than that of any of the other elements of this group. Bonds between nitrogen and oxygen, therefore, will be considerably less polar than those between oxygen and phosphorus, or oxygen and arsenic, antimony, or bismuth. Consequently, for this reason alone, the covalent contribution to the nitrogen–oxygen bond energy will be relatively more important than is the case with the bonds between oxygen and the heavier elements of the group. Thus, single-bond weakening by the lone pair—and a corresponding tendency toward bond multiplicity—is likely to be much greater with oxides of nitrogen than with oxides of the heavier nitrogen group elements.

For a list of some of the chief properties of the nitrogen group elements, see Table 19.

INDIVIDUAL NITROGEN GROUP ELEMENTS

Nitrogen. *Occurrence and distribution.* The atmosphere of the Earth consists of 78.03 weight percent of nitrogen; this is the principal source of nitrogen for commerce and industry. The atmosphere also contains varying small amounts of ammonia and ammonium salts, as well as nitrogen oxides and nitric acid (the latter substances being formed in electrical storms and in the internal combustion engine). Nitrogen occurs also in mineral deposits of nitre or saltpetre (potassium nitrate, KNO_3) and Chile saltpetre (sodium nitrate, $NaNO_3$), but these deposits exist in quantities that are wholly inadequate for man's needs. Another material rich in nitrogen is guano, found in bat caves and in dry places frequented by birds. Nitrogen constitutes on the average about 16 percent by weight of the complex organic compounds known as proteins, present in all living organisms. The natural abundance of nitrogen in the earth's crust is 0.3 parts per 1,000. The cosmic abundance—the estimated total abundance in the universe—is between three and seven atoms per atom of silicon, which is taken as the standard.

Commercial production and uses. Commercial production of nitrogen is largely by fractional distillation of liquefied air. The boiling temperature of nitrogen is −195.8° C, about 13° below that of oxygen, which is therefore left behind. Nitrogen can also be produced on a large scale by burning carbon or hydrocarbons in air and separating the resulting carbon dioxide and water from the residual nitrogen. Various laboratory reactions that yield nitrogen include heating ammonium nitrite (NH_4NO_2) solutions, oxidation of ammonia by bromine water, and oxidation of ammonia by hot cupric oxide.

$$NH_4^+ + NO_2^- \rightarrow N_2 + 2\ H_2O$$
$$8\ NH_3 + 3\ Br_2 \rightarrow N_2 + 6\ NH_4^+ + 6\ Br^-$$
$$2\ NH_3 + 3\ CuO \rightarrow N_2 + 3\ H_2O + 3\ Cu$$

Uses of molecular nitrogen

Elemental nitrogen can be used as an inert atmosphere for reactions requiring the exclusion of oxygen. In the liquid state, nitrogen has valuable cryogenic applications; except for the gases hydrogen, methane, carbon monoxide, fluorine, and oxygen, practically all chemical substances have negligible vapour tensions at the boiling point of nitrogen and exist, therefore, as crystalline solids at that temperature.

In the chemical industry, nitrogen is used as a preventive of oxidation or other deterioration of a product, as an inert diluent of a reactive gas, as a carrier to remove heat or chemicals and as an inhibitor of fire or explosions. In the food industry nitrogen gas is employed to prevent spoilage through oxidation, mold, or insects, and liquid nitrogen is used for freeze drying and for refrigeration systems. In the electrical industry nitrogen is used to prevent oxidation and other chemical reactions, to pressurize cable jackets, and to shield motors. Nitrogen finds application in the metals industry in welding, soldering, and brazing, where it helps prevent oxidation, carburization, and decarburization. As a nonreactive gas, nitrogen is employed to make foamed—or expanded—rubber, plastics, and elastomers, to serve as a propellant gas for aerosol cans, and to pressurize liquid propellants for reaction jets. In medicine rapid freezing with liquid nitrogen may be used to preserve blood, bone marrow, tissue, bacteria, and semen.

Nitrogen fixation

Although the other applications are important, by far the greatest bulk of elemental nitrogen is consumed in the manufacture of nitrogen compounds. The bond between atoms in the nitrogen molecules is so strong (226 kilocalories per mole; more than twice that of molecular hydrogen) that it is difficult to cause molecular nitrogen to enter into other combinations. Most living organisms cannot utilize nitrogen directly and must have access to its compounds. Therefore the fixation of nitrogen, the incorporation of elemental nitrogen into compounds, is vitally important. In nature, two principal processes of nitrogen fixation are known. One is the action of electrical energy on the atmosphere, which dissociates nitrogen and oxygen molecules, allowing the free atoms to form nitric oxide, NO, and nitrogen dioxide, NO_2. Nitrogen dioxide then reacts with water as follows:

$$3\ NO_2 + H_2O \rightarrow 2\ HNO_3 + NO.$$

The nitric acid, HNO_3, dissolves and comes to Earth with rain as a very dilute solution. In time it becomes part of the combined nitrogen of the soil. The other principal process of natural nitrogen fixation is that of certain plants and vegetables called legumes. Through a cooperative action with bacteria, legumes are able to convert atmospheric nitrogen directly into nitrogen compounds. Certain bacteria alone, such as *Azotobacter chroococcum* and *Clostridium pasteurianum,* are also capable of fixing nitrogen.

The chief commercial method of fixing nitrogen is the Haber process for synthesizing ammonia. This process was developed during World War I to lessen the dependence of Germany on Chilean nitrate. It involves the direct synthesis of ammonia from its elements.

Properties and reaction. Nitrogen is a colourless, odour-

Table 19: Some Properties of the Nitrogen Group Elements

	nitrogen	phosphorus (white)	arsenic	antimony	bismuth
Atomic number	7	15	33	51	83
Atomic weight	14.0067	30.9738	74.9216	121.75	208.980
Colour of element	colourless	white	steel gray	silver	pinkish silver
Melting point (° C)	—209.86	44.1	817 (28 atm)	630.5	271.3
Boiling point (° C)	—195.8	280	613 (sublimes)	1,380	1,560
Density at 25° C (g/cm³)	1.25 (g/l)	1.82	5.73	6.684	9.8
Solubility in water (vol/vol water)	0.0231	none	none	none	none
Valence	3, (5)	3, 5	3, 5	3, 5	3, 5
Electronic configuration	$1s^22s^22p^3$	(Ne)$3s^23p^3$	(Ar)$3d^{10}4s^24p^3$	(Kr)$4d^{10}5s^25p^3$	(Xe)$4f^{14}5d^{10}6s^26p^3$
Isotopic abundance (terrestrial, percent)	^{14}N (99.63), ^{15}N (0.37)	^{31}P (100)	^{75}As (100)	^{121}Sb (57.25), ^{123}Sb (42.75)	^{209}Bi (100)
Radioactive isotopes (mass numbers)	12, 13, 16, 17	28–30, 32–34	68–74, 76–81, 85	112–120, 122, 124–135	196, 199–208, 210–215
Heat of fusion (cal/g)	6.1	5.03	88.5	38.3	12.5
Heat of vaporization (cal/g)	47.7	130	102	161	204.3
Specific heat (cal/g/° C)	0.249 (N_2)	0.189	0.082	0.0494	0.0294
Critical temperature (° C)	—147.1	675	—	—	—
Critical pressure (atm)	33.5	80	—	—	—
Electrical resistivity (microhm-cm)	—	1×10^{17}	33.3 (20° C)	39 (0° C)	106.8 (0° C)
Hardness (Mohs scale)	—	0.5	3.5	3.0–3.5	2.5
Crystal structure at 20° C	—	cubic	hexagonal	hexagonal	hexagonal
Radius					
Covalent (Å)	0.74	1.10	1.19	1.38	1.46
Ionic (M^{3+}, Å)	0.16	0.44	0.58	0.76	0.96
Ionization energy (first, kcal/mole)	337	255	228	201	170
Electronegativity					
(Pauling)	3.05	2.15	2.10	2.05	1.8
(Sanderson)	4.49	3.34	3.91	3.37	3.16

less gas, which condenses at −195.8° C to a colourless, mobile liquid. The element exists as N_2 molecules, represented as :N:::N:, for which the bond energy of 226 kilocalories per mole is exceeded only by that of carbon monoxide, 256 kilocalories per mole. Because of this high bond energy the activation energy for reaction of molecular nitrogen is usually very high, causing nitrogen to be relatively inert to most reagents under ordinary conditions. Furthermore, the high stability of the nitrogen molecule contributes significantly to the thermodynamic instability of many nitrogen compounds, in which the bonds, although reasonably strong, are far less so than those in molecular nitrogen. For these reasons, elemental nitrogen appears to conceal quite effectively the truly reactive nature of its individual atoms.

A relatively recent and unexpected discovery is that nitrogen molecules are able to serve as ligands in complex coordination compounds. The observation that certain solutions of ruthenium complexes can absorb atmospheric nitrogen has led to hope that one day a simpler and better method of nitrogen fixation may be found.

Atomic nitrogen

An active form of nitrogen, presumably containing free nitrogen atoms, can be created by passage of nitrogen gas at low pressure through a high-tension electrical discharge. The product glows with a yellow light and is much more reactive than ordinary molecular nitrogen, combining with atomic hydrogen and with sulfur, phosphorus, and various metals, and capable of decomposing nitric oxide, NO, to N_2 and O_2.

A nitrogen atom has the electronic structure represented by $1s^2$, $2s^2$, $2p^3$, or alternatively, 2–5. The five outer shell electrons screen the nuclear charge quite poorly, with the result that the effective nuclear charge felt at the covalent radius distance is relatively high. Thus nitrogen atoms are relatively small in size and high in electronegativity, being intermediate between carbon and oxygen in both of these properties. The electronic configuration includes three half-filled outer orbitals, which give the atom the capacity to form three covalent bonds. The nitrogen atom should therefore be a very reactive species, combining with most other elements to form stable binary compounds, especially when the other element is sufficiently different in electronegativity to impart substantial polarity to the bonds. When the other element is lower in electronegativity than nitrogen, the polarity gives partial negative charge to the nitrogen atom, making its lone-pair electrons available for coordination. When the other element is more electronegative, however, the resulting partial positive charge on nitrogen greatly limits the donor properties of the molecule. When the bond polarity is low (due to the electronegativity of the other element being similar to that of nitrogen), multiple bonding is greatly favoured over single bonding. If disparity of atomic size prevents such multiple bonding, then the single bond that forms is likely to be relatively weak, and the compound is likely to be unstable with respect to the free elements. All of these bonding characteristics of nitrogen are observable in its general chemistry.

Analytical chemistry. Often the percentage of nitrogen in gas mixtures can be determined by measuring the volume after all other components have been absorbed by chemical reagents. Decomposition of nitrates by sulfuric acid in the presence of mercury liberates nitric oxide, which can be measured as a gas. Nitrogen is released from organic compounds when they are burned over copper oxide, and the free nitrogen can be measured as a gas after other combustion products have been absorbed. The well-known Kjeldahl method for determining the nitrogen content of organic compounds involves digestion of the compound with concentrated sulfuric acid (optionally containing mercury, or its oxide, and various salts, depending on the nature of the nitrogen compound). In this way, the nitrogen present is converted to ammonium sulfate. Addition of an excess of sodium hydroxide releases free ammonia, which is collected in standard acid; the amount of residual acid, which has not reacted with ammonia, is then determined by titration.

The Kjeldahl method

Biological and physiological significance. As might be expected in view of the importance of the presence of nitrogen in living matter, most—if not all—organic nitrogen compounds are physiologically active.

The bends

Nitrogen itself, being inert, is innocuous except when breathed under pressure, in which case it dissolves in the blood and other body fluids in higher than normal concentration. This in itself produces a narcotic effect, but if the pressure is reduced too rapidly, the excess nitrogen evolves as bubbles of gas in various locations in the body. These can cause muscle and joint pain, fainting, partial paralysis, and even death. These symptoms are referred to as "the bends." Divers, and others forced to breathe air under pressure, must therefore be extremely careful that the pressure is reduced to normal very slowly following exposure. This enables the excess nitrogen to be released harmlessly through the lungs without forming bubbles. A better alternative is to substitute mixtures of oxygen and helium for air. Helium is much less soluble in body fluids, and the dangers are thus diminished.

Isotopes of nitrogen. Nitrogen exists as two stable isotopes, ^{14}N and ^{15}N. These can be separated by chemical exchange or by thermal diffusion. Artificial radioactive isotopes have masses of 12, 13, 16, and 17. The most stable has a half-life of only about 10 minutes.

Phosphorus. *Occurrence and distribution.* Phosphorus is a very widely distributed element—12th most abundant in the Earth's crust, to which it contributes about 0.10 weight percent. Its cosmic abundance is estimated to be about one atom per 100 atoms of silicon, the standard. Its high chemical reactivity assures that it does not occur in the free state. The principal combined forms in nature are the phosphate salts. Nearly 190 different minerals have been found to contain phosphorus, but, of these, the principal source of phosphorus is the apatite series in which calcium ions exist along with phosphate ions and variable amounts of fluoride, chloride, or hydroxide ions, according to the formula $[Ca_{10}(PO_4)_6(F, Cl, \text{or } OH)_2]$. Commonly such metal atoms as magnesium, manganese, strontium, and lead substitute for calcium in the mineral; and silicate, sulfate, vanadate, and similar anions substitute for phosphate ions. Very large sedimentary deposits of fluoroapatite are found in many parts of the Earth. The phosphate of bone and tooth enamel is hydroxyapatite. (The principle of lessening tooth decay by fluoridation depends upon the conversion of hydroxyapatite to the harder, more decay resistant, fluoroapatite.)

Estimates of the total phosphate rock in the Earth's crust average about 50,000,000,000 tons, of which north Africa contains two-thirds, and the U.S.S.R. and the United States most of the remaining third. This estimate includes only ore that is sufficiently rich in phosphate for conversion to useful products by present methods. Vast quantities of material lower in phosphorus content also exist.

Acidulation and reduction

Commercial production and uses. Two principal techniques for converting phosphate rock to usable materials are practiced. One involves acidulation of the crushed rock—with either sulfuric or phosphoric acids—to form crude calcium hydrogen phosphates that, being water soluble, are valuable additions to fertilizer. The other method is the reduction of the phosphate with carbon in an electric furnace to give elemental phosphorus. The latter reaction is extremely complex, and its precise details depend upon the composition of the mineral phosphate. A charge of sand, coke, and phosphate rock is melted at about 1,500° C in an electric furnace. The calcium and impurities are left in the form of a complex fluorosilicate slag, and elemental phosphorus vapour, at about 300° C, distills out and is collected, condensed, and stored underwater as the white allotropic form of the element. More than half a million tons of phosphorus are made annually in the United States in this way. Most of the output is burned to phosphoric anhydride and subsequently treated with water to form phosphoric acid, H_3PO_4.

Only about 5 percent of the 2,000,000 tons of phosphorus consumed per year in the United States is used in the elemental form. Pyrotechnic applications of the element include tracers, incendiaries, fireworks, and matches. Some is used as an alloying agent, some to kill rodents, and the rest is employed in chemical synthesis. A large amount is converted to sulfides used in matches and in the manufacture of insecticides and oil additives. Most of the remainder is converted to halides or oxides for subsequent use in synthesizing organic phosphorus compounds.

Properties and reactions. The electronic configuration of the phosphorus atom can be represented by $1s^2$, $2s^2$, $2p^6$, $3s^2$, $3p^3$, or 2–8–5. The outer shell arrangement therefore resembles that of nitrogen, with three half-filled orbitals each capable of forming a single covalent bond and an additional lone-pair of electrons. Depending on the electronegativity of the elements with which it combines, phosphorus can therefore exhibit oxidation states of +3 or −3, just as does nitrogen. The principal differences between nitrogen and phosphorus are that the latter is of considerably lower electronegativity and has larger atoms, with outer *d* orbitals available. For these reasons, the similarities between nitrogen and phosphorus chemistry are largely formal ones, tending to conceal the actual, wide differences. The outer *d* orbitals in phosphorus permit an expansion of the octet, which leads to the +5 state, with five actual covalent bonds being formed in compounds, a condition impossible for nitrogen to achieve.

Allotropic forms

The first striking difference in chemistry of the two elements is that elemental phosphorus exists under ordinary conditions in any of several modifications, or allotropic forms, all of which are solid. Phosphorus molecules of formula P_2, structurally analogous to N_2 molecules and evidently also triply bonded, exist only at very high temperatures. These P_2 molecules do not persist at lower temperatures—below about 800°—because of the fact that three single bonds in phosphorus, in contrast to the situation with nitrogen, are favoured over one triple bond. On cooling, the triply bonded P_2 molecules condense to form tetrahedral P_4 molecules, in which each atom is joined to three others by single bonds. These molecules further condense to form either hexagonal- or cubic-structured molecular solids, both called "white phosphorus." Because of the relatively weak intermolecular attractions (van der Waals forces) between the separate P_4 molecules, the solid melts easily at 44.1° C and boils at about 280° C. Formation of tetrahedra requires bond angles of 60° instead of the preferred 90°–109° angles, so that white phosphorus is a relatively unstable, or metastable, form. It changes spontaneously, but slowly, at temperatures around 200° or higher, to a polymeric form called "red phosphorus." This substance is amorphous when formed at lower temperatures, but it can become crystalline, with a melting point of about 590° C. At higher temperatures and pressures, or with the aid of a catalyst, at ordinary pressures and a temperature of about 200° C, phosphorus is converted to a black crystalline form, which somewhat resembles graphite. This may prove to be the most stable form of phosphorus, despite the relative difficulty in its preparation. In both the red and the black forms, each phosphorus atom forms three single bonds, which are spread apart sufficiently to be relatively strain free.

Consistent with the metastable condition of the white modification, and the crowding of its covalent bonds, this form is far more reactive chemically than the others. It is highly toxic, reacts vigorously with most reagents, and inflames in air at only 35° C, so that it must be stored under water or other inert liquid. White phosphorus dissolves readily in solvents such as carbon disulfide, in which it maintains the composition P_4. In contrast, red phosphorus is insoluble and relatively inert, although large quantities of the usual commercial form can ignite spontaneously in air and react with water to form phosphine and phosphorus oxyacids. Black phosphorus is more inert and is capable of conducting electricity. Both these polymeric forms are insoluble and are very much less volatile than white phosphorus.

Analytical chemistry. Elemental phosphorus can be detected by its phosphorescence. It can also be converted to phosphine with boiling sodium hydroxide solution or with zinc and sulfuric acid; the phosphine is identified by means of test paper containing either silver nitrate or mercuric chloride, both of which are reduced to the free metal by phosphine, thereby darkening the paper. Phosphorus vapour also readily darkens silver nitrate test paper. Phosphorus is determined quantitatively by oxidation to phosphate, followed by any of several standard procedures. Phosphate, for example, may be precipitated as the magnesium ammonium salt, $MgNH_4PO_4$, converted by ignition to magnesium pyrophosphate, $Mg_2P_2O_7$, and weighed. Alternatively, phosphate may be precipitated as ammonium phosphomolybdate; this can be weighed as such, converted to magnesium pyrophosphate and weighed, or titrated directly with sodium hydroxide solution.

Biological and physiological significance. Phosphorus is an important constituent of bones and teeth, and it is essential to the growth of living organisms. In organisms the element usually appears as phosphate. In its other forms phosphorus is likely to prove very toxic. White phosphorus attacks the skin and, when ingested, causes a necrosis of the jawbone, called "phossy jaw." Certain organic esters of phosphoric acid, used as lubricating-oil additives, have been found to cause permanent paralysis when accidentally ingested. Phosphine is extremely toxic, as are its organic derivatives. Some of the most toxic substances known to man, collectively termed nerve gas, are organic derivatives of phosphorus.

Nerve gas

Isotopes of phosphorus. The only naturally occurring isotope of phosphorus is that of mass 31. The other isotopes from mass 29 to mass 34 have been synthesized by appropriate nuclear reactions. All of these are radioactive with relatively short half-lives. The isotope of mass 32 has a half-life of about 14 days and has proven extremely useful in tracer studies involving the absorption and movement of phosphorus in living organisms.

Arsenic. *Occurrence and distribution.* The abundance of arsenic in the Earth's crust is about five grams per ton; the cosmic abundance is estimated as about four atoms per million atoms of silicon. The element is widely distributed. A small amount exists in the native state, in 90–98 percent purity. Most, however, is combined in more than 150 different minerals, as sulfides, arsenides, sulfoarsenides, and arsenites. Mispickel, or arsenopyrite, FeAsS, is among the most common of arsenic-bearing minerals; others are realgar, As_4S_4; orpiment, As_2S_3; loellingite, $FeAs_2$; and enargite, Cu_3AsS_4. Most commercial arsenic is recovered as a by-product of the smelting of copper, lead, cobalt, and gold ores.

Commercial production and uses. Metallic arsenic forms when arsenopyrite is heated at 650°–700° C in the absence of air. The arsenic in arsenopyrite and the arsenic impurities in other metal ores unite readily with oxygen when heated in air, forming the easily sublimed oxide, As_4O_6, also known as "white arsenic." The vapour of the oxide is collected and condensed in a series of brick chambers and later purified by resublimation. Most arsenic is prepared by carbon reduction of the arsenious oxide dust thus collected.

World consumption of metallic arsenic is relatively small, only a few hundred tons per year. Most of what is consumed comes from Sweden. It is used in metallurgical applications because of its metalloid properties. About one percent arsenic content is desirable in the manufacture of lead shot, for example, because it improves the roundness of the molten drops. Bearing alloys based on lead are improved in both thermal and mechanical properties when they contain about 3 percent arsenic. A small amount of arsenic in lead alloys hardens them for use in batteries and cable sheathing. Small concentrations of arsenic improve the corrosion resistance and thermal properties of copper and brass. Very highly purified arsenic finds applications in semiconductor technology, where it is used with silicon and germanium, as well as in the form of gallium arsenide, GaAs, for diodes, lasers, and transistors.

In contrast to the small use of metallic arsenic, tens of thousands of tons of the element are consumed annually in the form of its compounds. These are used primarily in agriculture.

Properties and reactions. In its most stable free state, arsenic is a steel-gray, brittle solid with low thermal and electrical conductivity. Other forms have been reported but are not well characterized, including especially a yellow, metastable form, which may consist of As_4 molecules analogous to white phosphorus, P_4. Arsenic sublimes at 613° C, and in the vapour it exists as As_4 molecules, which do not begin to dissociate until about 800° C; dissociation to As_2 molecules becomes complete at about 1,700° C.

Electronic structure

The electronic structure of the arsenic atom, 2–8–18–5, resembles those of nitrogen and phosphorus in that there are five electrons in the outermost shell, but it differs from them in having 18 electrons in the penultimate shell instead of two or eight. The addition of ten positive charges to the nucleus during the filling of the five $3d$ orbitals frequently causes a general contraction of the electronic cloud and a concomitant increase in electronegativity of the elements. In other groups of the periodic table this is clearly shown. Thus, it seems generally accepted that zinc is more electronegative than magnesium and, similarly, that gallium is more electronegative than aluminum. The difference diminishes, however, in the next groups, and many authorities do not agree that germanium is more electronegative than silicon, although an abundance of chemical evidence appears to indicate that this is so. The similar transition from penultimate 8-shell to 18-shell element in passing from phosphorus to arsenic might also produce an increase in the electronegativity of arsenic over phosphorus, but this remains controversial.

Bonding properties

The outer-shell similarity of the two elements suggests that arsenic, like phosphorus, can form three covalent bonds per atom, with an additional lone pair of electrons left unbonded. The oxidation state of arsenic should, therefore, be either +3 or −3 depending on the electronegativity of arsenic and that of the elements with which it is combined. The possibility should also exist of utilizing the outer d orbitals to expand the octet, thereby allowing arsenic to form five bonds. This possibility is realized only in compounds with fluorine. The availability of the lone pair for complex formation (through electron donation) appears much less in the arsenic atom than in phosphorus and nitrogen, as evidenced by the chemistry of the element.

Arsenic itself is stable in dry air, but in moist air it tends to become coated with a black oxide. Sublimed arsenic vapour readily burns in air to form arsenious oxide. The free element is essentially unaffected by water, bases, or nonoxidizing acids, but it can be oxidized by nitric acid to the +5 state. Halogens attack arsenic, as does sulfur, and the element will combine directly with many metals forming arsenides.

Tests for arsine

Analytical chemistry. Qualitatively, arsenic may be detected by precipitation as the yellow arsenious sulfide from hydrochloric acid of 25 percent or greater concentration. Trace amounts of arsenic are usually determined by conversion to arsine. The latter can be detected by the so-called Marsh test, in which arsine is thermally decomposed, forming a black arsenic mirror inside a narrow tube, or by the Gutzeit method, in which a test paper impregnated with mercuric chloride darkens when exposed to arsine because of the formation of free mercury.

Biological and physiological significance. The toxicity of arsenic and its compounds varies widely, ranging from the exceedingly poisonous arsine—and its organic derivatives—to elemental arsenic itself, which is relatively inert. Arsenical compounds in general are skin irritants, which easily cause dermatitis. Protection against inhalation of arsenic-containing dusts is recommended, but most poisoning appears to come from ingestion. The maximum tolerable concentration of arsenic in dusts during an eight-hour day is 0.5 milligrams per cubic metre. For arsine, exposure of similar duration requires that the concentration be less than 0.05 parts per million in the air. In addition to the many uses of arsenic compounds as herbicides and pesticides, they have in several instances been employed as pharmacological agents. The first successful antisyphilitic agent, for example, was an arsenic compound, "Salvarsan," or "606," or 3,3′-diamino-4,4′-dihydroxyarsenobenzene dihydrochloride.

Isotopes of arsenic. Only one stable isotope of arsenic, that of mass 75, occurs in nature. Among the artificial radioactive isotopes is one of mass 76, which has a half-life of 26.4 hours.

Antimony. *Occurrence and distribution.* Antimony is about one-fifth as abundant as arsenic, contributing on the average about one gram to every ton of the Earth's crust. Its cosmic abundance is estimated as about one atom to every 5,000,000 atoms of silicon. Small deposits of native metal have been found, but most of the antimony occurs in the form of more than 100 different minerals. The most important of these is stibnite, Sb_2S_3. Small stibnite deposits are found in Algeria, Bolivia, China, Yugoslavia, Mexico, Peru, and South Africa. Some economic value also attaches to kermesite ($2Sb_2S_3 \cdot Sb_2O_3$), argentiferous tetrahedrite [$(Cu,Fe)_{12}Sb_4S_{13}$], livingstonite ($HgSb_4S_7$), and jamesonite ($Pb_4FeSb_6S_{14}$). Small amounts are also recoverable from the production of copper and lead. About half of all the antimony produced is reclaimed from scrap lead alloy from old batteries, to which antimony had been added to provide hardness.

Stibnite

Commercial production and uses. High-grade or enriched stibnite reacts directly with scrap iron in the molten state, liberating antimony metal. The metal can also be obtained by conversion of stibnite to the oxide, followed by reduction with carbon. Sodium sulfide solutions are effective leaching agents for the concentration of stibnite from ores. Electrolysis of these solutions produces antimony. After further purification of the crude antimony, the metal, called regulus, is cast into cakes.

About half of this antimony is used metallurgically, principally in alloys. It improves the hardness and corrosion resistance of lead. Most of the metal is used for lead alloys, largely for storage batteries, but also for chemical equipment such as tanks, pipes, and pumps. With tin, antimony forms such alloys as britannia metal and pewter, used for utensils and Babbitt metal for bearings. Other applications of antimony alloys are in solder, type metal, and other special materials. Highly purified antimony is used in semiconductor technology to prepare the intermetallic compounds indium, aluminum, and gallium antimonide for diodes and infrared detectors.

Properties and reactions. The most stable form of antimony is a brittle, silvery solid of high metallic lustre. Electrolytic deposition of antimony under certain conditions produces an unstable, amorphous form called "explosive antimony," because, when bent or scratched, it will change in a mildly explosive manner to the more stable, metallic form. There is also an amorphous black form of antimony that results from sudden quenching of the vapour, and a yellow form produced by low temperature oxidation of stibine, SbH_3, with air or chlorine. Metallic antimony is not affected by air or moisture under ordinary conditions, but it can be oxidized easily by oxygen, sulfur, and the halogens, especially when heated.

Electronic structure

The electronic structure of antimony closely resembles that of arsenic. It is represented as 2–8–18–18–5, with three half-filled orbitals in the outermost shell. Thus it can form three covalent bonds and exhibit +3 and −3 oxidation states. The electronegativity of antimony, like that of arsenic, remains somewhat controversial. It is generally agreed to be lower than that of arsenic, but whether it is lower also than that of phosphorus is undecided. It can act as an oxidizing agent and reacts with many metals to form antimonides that, in general, resemble nitrides, phosphides, and arsenides, but are somewhat more metallic. The promotion of one of the lone-pair electrons to an outer d orbital apparently occurs more easily with antimony than with arsenic, since antimony exhibits the +5 oxidation state in forming both the pentafluoride and the pentachloride.

Analytical chemistry. Antimony may be separated and weighed for analysis as the sulfide, Sb_2S_3. Alternatively, the sulfide may be converted to the oxide and, after careful ignition, weighed as Sb_4O_6. Numerous volumetric methods are also available, including several methods of oxidizing Sb(III) with potassium permanganate, potassium bromate, or iodine. In the absence of arsenic, small amounts of antimony may be determined by a modified Gutzeit method.

Biological and physiological significance. Antimony and a number of its compounds are highly toxic. In fact, the use of antimony compounds for medicinal purposes was temporarily outlawed several centuries ago because of the number of fatalities they had caused. A hydrated potassium antimonyl tartrate called "tartar emetic" is currently used in medicine as an expectorant, diaphoretic,

Tartar emetic

and emetic. The maximum tolerable concentration of antimony dust in air is about the same as for arsenic, 0.5 milligrams per cubic metre.

Isotopes of antimony. Two stable isotopes, nearly equal in abundance, occur in nature. One has mass 121 and the other mass 123. Radioactive isotopes of masses 120, 122, 124, 125, 126, 127, 129, and 132 have been prepared.

Bismuth. *Occurrence and distribution.* Bismuth is about as abundant as silver, contributing about 2×10^{-5} weight percent of the Earth's crust. Its cosmic abundance is estimated as about one atom to every 7,000,000 atoms of silicon. It occurs both native and in compounds. In the native state, it is found in veins associated with lead, zinc, tin, and silver ores in Bolivia, Canada, England, and Germany. Its naturally occurring compounds are chiefly the oxide (bismite or bismuth ochre), the sulfide (bismuthinite or bismuth glance), and two carbonates (bismutite and bismutosphaerite). Commercial bismuth, however, is produced largely as a by-product in the smelting and refining of lead, tin, copper, silver, and gold ores. Thus, it comes—for example—from tungsten ores in South Korea, lead ores in Mexico, copper ores in Bolivia, and both lead and copper ores in Japan.

Commercial production and uses. Bismuth is volatile at high temperature, but it usually remains with the other metals after smelting operations. Electrolytic refining of copper leaves bismuth behind as one component of the anode sludge. Separation of bismuth from lead by the Betterton–Kroll process involves the formation of highmelting calcium or magnesium bismuthide (Ca_3Bi_2 or Mg_3Bi_2), which separates and can be skimmed off as dross. The dross may be chlorinated to remove the magnesium or calcium, and finally the entrained lead. Caustic soda treatment then produces highly pure bismuth. An alternative separation, the Betts process, involves electrolytic refining of lead bullion (containing bismuth and other impurities) in a solution of lead fluosilicate and free fluosilicic acid, bismuth being recovered from the anode sludge. Separation of bismuth from its oxide or carbonate ores can be effected by leaching with concentrated hydrochloric acid. Dilution then precipitates the oxychloride, BiOCl. This—heated with lime and charcoal—produces metallic bismuth.

Alloys of bismuth

Metallic bismuth is used principally in alloys, to many of which it imparts its own special properties of low melting point and expansion on solidification. Bismuth is thus a useful component of type-metal alloys, which make neat, clean castings; and it is an important ingredient of lowmelting alloys, called fusible alloys, which have a large variety of applications, especially in fire-detection equipment. A bismuth–manganese alloy has been found effective as a permanent magnet. Small concentrations of bismuth improve the machinability of aluminum, steel, stainless steels, and other alloys and suppress the separation of graphite from malleable cast iron. Thermoelectric devices for refrigeration make use of bismuth telluride, Bi_2Te_3, and bismuth selenide, Bi_2Se_3. Liquid bismuth has been used as a fuel carrier and coolant in the generation of nuclear energy.

The principal chemical application of bismuth is in the form of bismuth phosphomolybdate, which is an effective catalyst for the air oxidation of propylene and ammonia to acrylonitrile. The latter is used to make acrylic fibres, paints, and plastics. Pharmaceutical uses of bismuth have been practiced for centuries. It is effective in indigestion remedies and antisyphilitic drugs. Slightly soluble or insoluble salts are utilized in the treatment of wounds and gastric disorders, and bismuth is sometimes injected in the form of finely divided metal, or as suspensions of its insoluble salts. Substantial quantities of the oxychloride, BiOCl, have been used to impart a pearlescent quality to lipstick, nail polish, and eye shadow.

Properties and reactions. Bismuth is a rather brittle metal with a somewhat pinkish, silvery metallic lustre. It undergoes a 3.3 percent expansion when it solidifies from the molten state. Its electrical conductivity is very poor, but somewhat better in the liquid state than in the solid. With respect to thermal conductivity, it is the poorest of all metals except mercury. Bismuth is quite resistant to corrosion by air and moisture, but it is oxidized rapidly at its boiling point of 1,560° C. It is oxidized and dissolved by concentrated nitric acid.

Bismuth atoms have the same electronic structure in their outermost shell as do the other elements of the nitrogen group. They can, therefore, form three single covalent bonds, exhibiting either a +3 or −3 oxidation state. The element has a somewhat lower electronegativity than the others, and its lone pair of electrons is evidently quite inert, causing the +5 state of bismuth to be rare and unstable.

Analytical and physiological chemistry. Bismuth is usually determined gravimetrically, being precipitated and weighed as the phosphate or the oxychloride, BiOCl. To produce the latter, a suitable amount of hydrochloric acid is added to a nitric acid solution containing the bismuth, and the resulting solution is poured into a large volume of water, causing the oxychloride to precipitate. Volumetric and colorimetric methods of determination are also available.

Bismuth is relatively nontoxic, the least so of the heavy metals. It is generally not an industrial hazard. Although bismuth and certain of its compounds find considerable therapeutic use, some authorities recommend that other remedies be substituted. Soluble inorganic bismuth compounds are toxic.

Isotopes of bismuth. Bismuth forms only one stable isotope, that of mass 209. A large number of radioactive isotopes are known, as shown in the table, most of them being very unstable. (R.T.S./Ed.)

Oxygen group elements

Oxygen, sulfur, selenium, tellurium, and polonium comprise the family referred to in the periodic classification of the chemical elements as Group VIa, or the chalcogens (see Figure 10). A relationship among the first three members of the group was recognized as early as 1829; tellurium was assigned its place by 1865, and the discovery of polonium in 1898 completed the group. By the early 1970s, the possibility of finding element 116, the next member of Group VIa, in nature appeared to be negligible.

Estimates of the proportions of the various kinds of atoms in the universe put oxygen fourth in abundance, after hydrogen, helium, and neon, but the importance of such a ranking is slight since hydrogen atoms account for almost 94 percent of the total and helium for most of the rest. About three atoms out of 10,000 are oxygen, but because the mass of an oxygen atom is approximately 16 times that of a hydrogen atom, oxygen constitutes a larger fraction of the mass of the universe, though still only about 0.5 percent. In the regions ordinarily accessible to man, however—*i.e.,* within a few kilometres of the surface of the Earth—oxygen is the most abundant element: in mass, it makes up about 20 percent of the air, about 46 percent of the solid crust of the Earth, and about 89 percent of the water.

Oxygen is represented by the chemical symbol O. In the air, oxygen exists mostly as molecules each made up of two atoms (O_2), although small amounts of ozone (O_3), in which three atoms of oxygen make up each molecule, are present in the atmosphere. Oxygen is a colourless, odourless, tasteless gas essential to living organisms, being taken up by animals, which convert it to carbon dioxide; plants, in turn, utilize carbon dioxide as a source of carbon and return the oxygen to the atmosphere. Oxygen forms compounds by reaction with practically any other element, as well as by reactions that displace elements from their combinations with each other; in many cases, these processes are accompanied by the evolution of heat and light and in such cases are called combustions.

Cosmic abundance of sulfur

In cosmic abundance, sulfur ranks ninth among the elements, accounting for only one atom of every 20,000–30,000. Sulfur occurs in the uncombined state as well as in combination with other elements in rocks and minerals that are widely distributed, although it is classified among the minor constituents of the Earth's crust, in which its proportion is estimated to be between 0.03 and 0.06 percent. On the basis of the finding that certain meteorites contain about 12 percent sulfur, it has been suggested that deeper layers of the Earth contain a much larger proportion. Seawater contains about 0.09 percent sulfur in the form of sulfate. The most important source is underground deposits of very pure sulfur present in domelike geologic structures where the sulfur is believed to have been formed by the action of bacteria upon the mineral anhydrite, in which sulfur is combined with oxygen and calcium. Deposits of sulfur in volcanic regions probably originated from gaseous hydrogen sulfide generated below the surface of the Earth and transformed into sulfur by reaction with the oxygen in the air.

Sulfur exists under ordinary conditions as a pale yellow, crystalline, nonmetallic solid; it is odourless and tasteless, combustible, and insoluble in water. Its chemical symbol is S. It reacts with all metals except gold and platinum, forming sulfides; it also forms compounds with several of the nonmetallic elements. Several million tons of sulfur are produced each year, mostly for the manufacture of sulfuric acid, which is widely used in industry.

The element selenium (symbol Se) is much rarer than oxygen or sulfur, comprising approximately 90 parts per billion of the crust of the Earth. It is occasionally found uncombined, accompanying native sulfur, but is more often found in combination with heavy metals (as copper, mercury, lead, or silver) in a few minerals. The principal commercial source of selenium is as a by-product of copper refining; its major uses are in the manufacture of electronic equipment, in pigments, and in making glass. The gray, metallic form of the element is the most stable under ordinary conditions; this form has the unusual property of greatly increasing in electrical conductivity when exposed to light. Selenium compounds are toxic to animals; plants grown in seleniferous soils may concentrate the element and become poisonous.

Tellurium is a silvery-white element (symbol Te) with properties intermediate between those of metals and nonmetals; it makes up approximately one part per billion of the Earth's crust. Like selenium, it is less often found uncombined than as compounds of metals such as copper, lead, silver, or gold, and is obtained chiefly as a by-product of the refining of copper or lead. No large use for tellurium has been found.

Polonium (symbol Po) is an extremely rare, radioactive element found in minerals containing uranium. It has some scientific applications as a source of alpha radiation.

HISTORY

Oxygen. Although several substances now recognized as elements were known in their uncombined states in ancient times, the presently accepted idea of chemical elements dates from 1661, when the English natural philosopher Robert Boyle set forth their nature as simple, primitive, perfectly unmixed bodies, not formed from each other nor from any other bodies. Despite its terrestrial abundance, oxygen was not recognized as belonging to this class before the 1770s. Two chemists, Carl Wilhelm Scheele, in about 1772, and Joseph Priestley, in 1774, both obtained oxygen by heating certain metal oxides. Antoine-Laurent Lavoisier, with remarkable insight, interpreted the role of oxygen in respiration as well as combustion, discarding the phlogiston theory, which had been accepted up to that time; he noted its tendency to form acids by combining with many different substances and accordingly named the element *oxygen* (*oxygène*) from the Greek words for "acid former."

Boyle's definition of chemical elements

Sulfur. The history of sulfur is part of antiquity. The name itself probably found its way into Latin from the language of the Oscans, an ancient people who inhabited the region including Vesuvius, where sulfur deposits are widespread. Prehistoric man used sulfur as a pigment for cave painting; one of the first recorded instances of the art of medication is in the use of sulfur as a tonic.

The combustion of sulfur had a role in Egyptian religious ceremonials as long as 4,000 years ago. "Fire and brimstone" references in the Bible are related to sulfur, suggesting that "hell's fires" are fuelled by sulfur. The beginnings of practical and industrial uses of sulfur are credited to the Egyptians, who used sulfur dioxide for bleaching cotton as early as 1600 BC. Greek mythology includes sulfur chemistry: Homer tells of Odysseus' use of sulfur dioxide to fumigate a chamber in which he had slain his wife's suitors. The use of sulfur in explosives and fire displays dates to about 500 BC in China, and flameproducing agents used in warfare (Greek fire) were prepared with sulfur in the Middle Ages. Pliny the Elder in AD 50 reported a number of individual uses of sulfur and ironically was himself killed, in all probability by sulfur fumes, at the time of the great Vesuvius eruption (AD 79). Sulfur was regarded by the alchemists as the principle of combustibility. Lavoisier recognized it as an element in 1777, although it was considered by some to be a compound of hydrogen and oxygen; its elemental nature was established by the French chemists Joseph Gay-Lussac and Louis Thenard.

Selenium. In 1817 a Swedish chemist named Jöns Jacob Berzelius noted a red substance resulting from sulfide ores from mines of Fahlun (Falun), Sweden. When this red material was investigated in the following year, it proved to be an element and was named after the Moon or the Moon goddess Selene. An ore of unusually high selenium content was discovered by Berzelius only days before he made his report to the scientific societies of the world on selenium. His sense of humour is evident in the name he gave the ore, *eucairite,* meaning "just in time."

Tellurium. The element tellurium was isolated before it was actually known to be an elemental species. About 1782 Franz Joseph Müller von Reichenstein, an Austrian mineralogist, worked with an ore referred to as German gold. From this ore he obtained a material that defied his attempts at analysis and was called by him *metallum problematicum.* In 1798 Martin Heinrich Klaproth confirmed Müller's observations and established the elemental nature of the substance. He named the element after man's "heavenly body" Tellus, or Earth.

Polonium. The fifth member of Group VIa, polonium, was the first to be found by taking advantage of the phenomenon of radioactivity, discovered by Henri Becquerel in the last part of the 19th century. Pierre and Marie Curie isolated the element while carrying out analyses on a uranium ore, pitchblende. The very intense radioactivity not attributable to uranium was ascribed to a new element, named by them after Mme Curie's homeland, Poland. The discovery was announced in July 1898.

Intense radioactivity of polonium

COMPARISON OF PROPERTIES

The elements belonging to Group VIa of the periodic table are characterized by electron configurations in which six electrons occupy the outermost shell. An atom having such an electronic structure tends to form a stable shell of eight electrons by adding two more, producing an ion that has a double negative charge. This tendency to form negatively charged ions, typical of nonmetallic elements, is quantitatively expressed in the properties of electronegativity (the assumption of partial negative charge when present in covalent combination) and electron affinity (the ability of a neutral atom to take up an electron, forming a negative ion). Both these properties decrease in intensity as the elements increase in atomic number and mass proceeding down column VIa of the periodic table. Oxygen has, except for fluorine, the highest electronegativity and electron affinity of any element; the values of these properties then decrease sharply for the remaining members of the group to the extent that tellurium and polonium are regarded as predominantly metallic in nature, tending to lose rather than gain electrons in compound formation.

As is the case within all groups of the table, the lightest element—the one of smallest atomic number—has extreme or exaggerated properties. Oxygen, because of the small size of its atom, the small number of electrons in its underlying shell, and the large number of protons in the nucleus relative to the atomic radius, has properties uniquely different from those of sulfur and the remaining chalcogens. Those elements behave in a reasonably predictable and periodic fashion, as will be shown.

Although even polonium exhibits the oxidation state −2 in forming a few binary compounds of the type MPo (in which M is a metal), the heavier chalcogens do not form the negative state readily, favouring positive states such as +2 and +4. All the elements in the group except oxygen may assume positive oxidation states, with the even values predominating, but the highest value, +6, is not a very stable one for the heaviest members. When this state is achieved, there is a strong driving force for the atom to return to a lower state, quite often to the elemental form. This tendency makes compounds containing Se(VI) and Te(VI) more powerful oxidizing agents than S(VI) compounds. Conversely, sulfides, selenides, and tellurides, in which the oxidation state is −2, are strong reducing agents, easily oxidized to the free elements.

Neither sulfur nor selenium, and most certainly not oxygen, forms purely ionic bonds to a nonmetal atom. Tellurium and polonium form a few compounds that are somewhat ionic; tellurium(IV) sulfate, $Te(SO_4)_2$, and polonium(II) sulfate, $PoSO_4$, are examples.

Another feature of the Group VIa elements that parallels trends generally shown in columns of the periodic table is the increasing stability of molecules having the composition $X(OH)_n$ as the size of the central atom, X, increases. There is no compound HO—O—OH, in which the central oxygen atom would have a positive oxidation state, a condition that it resists. The analogous sulfur compound HO—S—OH, although not known in the pure state, does have a few stable derivatives in the form of metal salts, the sulfoxylates. More highly hydroxylated compounds of sulfur, $S(OH)_4$ and $S(OH)_6$, also do not exist, not because of sulfur's resistance to a positive oxidation state but rather because of the high charge density of the S(IV) and S(VI) states (the large number of positive charges relative to the small diameter of the atom), which repels the electropositive hydrogen atoms, and the crowding that attends covalent bonding of six oxygen atoms to sulfur, favouring loss of water:

$$S(OH)_6 \longrightarrow (HO)_2SO_2 + 2H_2O.$$

orthosulfuric acid   sulfuric acid   water

As the size of the chalcogen atom increases, the stability of the hydroxylated compounds increases: the compound orthotelluric acid, $Te(OH)_6$, is capable of existence.

Catenation. One of the most unusual properties of this family of elements is that of catenation or the bonding of an atom to another identical atom. Although oxygen shows this property only in the existence of ozone, sulfur is second only to carbon in exhibiting this mode of combination; the chalcogens beyond sulfur show it to diminishing degrees, polonium having no tendency to catenate. This type of bonding is found in the many ring systems of sulfur and selenium as well as in long zigzag chain structures. Catenation also occurs in the sulfanes and the metal polysulfides, compounds that have the formulas H_2S_x and M_2S_x, in which x may take the values of 2, 3, 4, or more, and M represents a singly charged metal ion. In comparing the catenation of sulfur atoms with that of carbon atoms, it may be noted that the number of molecular species having $(-S-)_x$ structures is very large, as is that of the analogous hydrocarbon compounds $(-CH_2-)_x$. The analogy between molecules containing rings of sulfur atoms and cyclic hydrocarbons is limited because only S_6 and S_8 have sufficient stability to permit proper comparison to be made. The general similarity extends to molecules of the form $Z(-S-)_xZ$ and $Z(-CH_2-)_xZ$, which are represented by compounds in which Z is H, SO_3H, and CF_3.

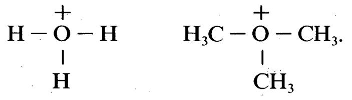
Covalent links between sulfur atoms have some of the character of multiple bonds—that is, more than one pair of electrons is shared, at least to some extent. Such interactions may involve overlap of p orbitals of one sulfur atom with d orbitals of another. Although not all investigators feel alike on the subject of d-orbital participation in the bonding of sulfur compounds, partial occupation of these orbitals is consistent with certain properties such as the colours of S_8 and S_2 molecules, the rigidity of chains and rings of sulfur atoms, and other features of the chemistry of sulfur compounds.

Manifestations of the similarities between sulfur and oxygen

Similarities of sulfur and oxygen are exhibited in certain compounds in which these elements interchange for one another. Examples include sulfates and thiosulfates (such as Na_2SO_4 and $Na_2S_2O_3$), phosphates and thiophosphates (containing the ions PO_4^{3-}, PO_3S^{3-}, $PO_2S_2^{3-}$, POS_3^{3-}, and PS_4^{3-}), and a similar series of arsenates and thioarsenates.

Ores of heavy metals often are found as both sulfides, MS, and selenides, MSe, or even with MS_xSe_y structures. The similarity in structures as well as properties accounts for the chalcogens' being found together in nature.

The number of atoms to which an element of Group VIa can form covalent bonds increases from oxygen to sulfur. An oxygen atom usually combines with two other atoms, as in the compounds water (H_2O), oxygen fluoride (OF_2), or dimethyl ether ($H_3C-O-CH_3$); the unshared pairs of electrons and the partial negative charge on the oxygen atom in most of these compounds allows bonding to another atom, as in the hydronium ion or trimethyloxonium ion:



$$\overset{+}{H-O-H} \quad \text{with } H \text{ below O} \qquad H_3C-\overset{+}{O}-CH_3. \text{ with } CH_3 \text{ below O}$$

hydronium ion   trimethyloxonium ion

Heavier members of the group associate or coordinate with other atoms or groups of atoms in numbers commensurate with the size of both the chalcogen and the coordinating group. Thus, sulfur tetrafluoride (SF_4) and sulfur hexafluoride (SF_6) are stable compounds, although sulfur hexaiodide (SI_6) is not known because of the very large size of the iodine atom. A closely related property is that of anionic complex formation: there is little evidence for the ion SF_6^{2-}, but there are ions such as $TeCl_6^{2-}$, TeF_6^{2-}, and PoI_6^{2-}.

Isotopes. The known isotopes of each of the Group VIa elements are listed in Table 20. Consistent with a generality observed throughout the periodic system, isotopes of even mass number are more abundant than those of odd mass number. Each member of the group except polonium has several stable isotopes; oxygen-18 and sulfur-35 have been used as tracers in chemical analysis, and polonium-210 serves as a convenient source of alpha particles (nuclei of helium atoms) for nuclear reactors and nuclear batteries.

INDIVIDUAL CHALCOGENS

Oxygen. *Natural occurrence and distribution.* As mentioned earlier, oxygen makes up about 46 percent of the mass of the crust of the Earth. In rocks, it is combined with metals and nonmetals in the form of oxides that are acidic (such as those of sulfur, carbon, aluminum, and phosphorus) or basic (such as those of calcium, magnesium, and iron) and as saltlike compounds that may be regarded as formed from the acidic and basic oxides, as sulfates, carbonates, silicates, aluminates, and phosphates. Plentiful as they are, these solid compounds are not useful as sources of oxygen, because separation of the element from its tight combinations with the metal atoms is too expensive.

Allotropy. Oxygen has two allotropic forms, diatomic (O_2) and triatomic (O_3, ozone). The properties of the diatomic form suggest that six electrons bond the atoms and two electrons remain unpaired, accounting for the paramagnetism of oxygen. The three atoms in the ozone molecule do not lie along a straight line.

Ozone may be produced from oxygen according to the equation:

$$3O_2 \longrightarrow 2O_3.$$

oxygen   ozone

Production of ozone

The process, as written, is endothermic (energy must be provided to make it proceed); conversion of ozone back into diatomic oxygen is promoted by the presence of transition metals or their oxides. Pure oxygen is partly transformed into ozone by a silent electrical discharge; the reaction is also brought about by absorption of ultraviolet light of wavelengths around 250 nanometres (nm, the nanometre, equal to 10^{-9} metre); occurrence of this process in the upper atmosphere removes radiation that

Table 20: Isotopes of the Group VIa Elements

	stable isotopes		unstable isotopes	
	mass	abundance (percentage)	mass	half-life
Oxygen	16	99.759	14	71 seconds
	17	0.037	15	2.07 minutes
	18	0.204	19	29 seconds
			20	14 seconds
Sulfur	32	95.0	30	1.4 seconds
	33	0.76	31	2.7 seconds
	34	4.22	35	88 days
	36	0.014	37	5.1 minutes
			38	2.78 hours
Selenium	74	0.87	70	44 minutes
	76	9.02	71	5 minutes
	77	7.58	72	8.4 days
	78	23.52	73†	42 minutes
	80	49.82		7.1 hours
	82	9.19	75	120 days
			77	17.5 seconds
			79	6.5 × 10^4 years
			81†	57 minutes
				18.6 minutes
			83†	70 seconds
				23 minutes
			84	3.3 minutes
			85	39 seconds
			87	16 seconds
Tellurium	120	0.089	114	16 minutes
	122	2.46	115	6 minutes
	123	0.87	116	2.5 hours
	124	4.61	117	61 minutes
	125	6.99	118	6 days
	126	18.71	119†	4.7 days
	128	31.79		16 hours
	130	34.48	121†	154 days
				17 days
			123	117 days
			125	58 days
			127†	109 days
				9.4 hours
			129†	34 days
				69 minutes
			131†	1.2 days
				25 minutes
			132	78 hours
			133†	50 minutes
				12.5 minutes
			134	42 minutes
Polonium*			192	0.5 second
			193	4 seconds
			194	0.5 second
			195	3 seconds
			196	6 seconds
			197	54 seconds
			198	1.7 minutes
			199	5 minutes
			200	11 minutes
			201	15.1 minutes
			202	45 minutes
			203	42 minutes
			204	3.6 hours
			205	1.8 hours
			206	8.8 days
			207†	2.8 seconds
				5.7 hours
			208	2.93 years
			209	103 years
			210	138.4 days
			211†	25 seconds
				0.52 second
			212	0.30 × 10^{-6} second
			213	4 × 10^{-6} second
			214	164 × 10^{-6} second
			215	0.0018 second
			216	0.15 second
			217	less than 10 seconds
			218	3.05 minutes

*No stable isotopes. †Two isotopes.

would be harmful to life on the surface of the Earth. The pungent odour of ozone is noticeable in confined areas in which there is sparking of electrical equipment, as in generator rooms. Ozone is light blue; its density is 1.658 times that of air, and it has a boiling point of −112° C at atmospheric pressure.

Ozone is a powerful oxidizing agent, capable of converting sulfur dioxide to sulfur trioxide, sulfides to sulfates, iodides to iodine (providing an analytical method for its estimation), and many organic compounds to oxygenated derivatives such as aldehydes and acids. The conversion by ozone of hydrocarbons from automotive exhaust gases to these acids and aldehydes contributes to the irritating nature of smog. Commercially, ozone has been used as a chemical reagent, as a disinfectant, in sewage treatment, water purification, and bleaching textiles.

Preparative methods. Production methods chosen for oxygen depend upon the quantity of the element desired. Laboratory procedures include the following:

1. Thermal decomposition of certain salts, such as potassium chlorate or potassium nitrate:

$$\begin{array}{ccccc} 2KClO_3 & \longrightarrow & 2KCl & + & 3O_2 \\ \text{potassium chlorate} & & \text{potassium chloride} & & \text{oxygen} \end{array}$$

$$\begin{array}{ccccc} 2KNO_3 & \longrightarrow & 2KNO_2 & + & O_2. \\ \text{potassium nitrate} & & \text{potassium nitrite} & & \text{oxygen} \end{array}$$

The decomposition of potassium chlorate is catalyzed by oxides of transition metals; manganese dioxide (pyrolusite,-MnO_2) is frequently used. The temperature necessary to effect the evolution of oxygen is reduced from 400° C to 250° by the catalyst.

2. Thermal decomposition of oxides of heavy metals:

$$\begin{array}{ccccc} 2HgO & \longrightarrow & 2Hg & + & O_2 \\ \text{mercury(II) oxide} & & \text{mercury} & & \text{oxygen} \end{array}$$

$$\begin{array}{ccccc} 2Ag_2O & \longrightarrow & 4Ag & + & O_2. \\ \text{silver(I) oxide} & & \text{silver} & & \text{oxygen} \end{array}$$

Scheele and Priestley used mercury(II) oxide in their preparations of oxygen.

3. Thermal decomposition of metal peroxides or of hydrogen peroxide:

$$\begin{array}{ccccc} 2BaO & + & O_2 & \xrightarrow{\text{heat}} & 2BaO_2 \\ \text{barium oxide} & & \text{oxygen (from air)} & & \text{barium peroxide} \end{array}$$

$$\begin{array}{ccccc} 2BaO_2 & \xrightarrow{\text{heat}} & 2BaO & + & O_2 \\ \text{barium peroxide} & & \text{barium oxide} & & \text{oxygen} \end{array}$$

$$\begin{array}{ccccccc} BaO_2 & + & H_2SO_4 & \longrightarrow & H_2O_2 & + & BaSo_4 \\ \text{barium peroxide} & & \text{sulfuric acid} & & \text{hydrogen peroxide} & & \text{barium sulfate} \end{array}$$

$$\begin{array}{ccccc} 2H_2O_2 & \xrightarrow[\text{catalyst}]{\text{heat or}} & 2H_2O & + & O_2 \\ \text{hydrogen peroxide} & & \text{water} & & \text{oxygen} \end{array}$$

An early commercial procedure for isolating oxygen from the atmosphere or for manufacture of hydrogen peroxide depended on the formation of barium peroxide from the oxide as shown in the equations.

4. Electrolysis of water containing small proportions of salts or acids to allow conduction of the electric current:

$$\begin{array}{ccccc} 2H_2O & \xrightarrow[\text{electric current}]{H_2SO_4} & 2H_2 & + & O_2. \\ \text{water} & & \text{hydrogen} & & \text{oxygen} \end{array}$$

Oxygen from the air

Commercial production and use. When required in tonnage quantities, oxygen is prepared by the fractional distillation of liquid air. The process takes advantage of the fact that when a compressed gas is allowed to expand, it cools. Major steps in the operation include the following: (1) Air is filtered to remove particulates; (2) moisture and carbon dioxide are removed by absorption in alkali; (3) the air is compressed and the heat of compression removed by ordinary cooling procedures; (4) the compressed and cooled air is passed into coils contained in a chamber; (5)

a portion of the compressed air (at about 200 atmospheres pressure) is allowed to expand in the chamber, cooling the coils; (6) the expanded gas is returned to the compressor with multiple subsequent expansion and compression steps resulting finally in liquefaction of the compressed air at a temperature of −196° C; (7) the liquid air is allowed to warm to distill first the light rare gases, then the nitrogen, leaving liquid oxygen. Multiple fractionations will produce a product pure enough (99.5 percent) for most industrial purposes.

The steel industry is the largest consumer of pure oxygen in "blowing" high carbon steel—that is, volatilizing carbon dioxide and other nonmetal impurities in a more rapid and more easily controlled process than if air were used. The treatment of sewage by oxygen holds promise for more efficient treatment of liquid effluents than other chemical processes. Incineration of wastes in closed systems using pure oxygen has become important. The so-called LOX of rocket oxidizer fuels is liquid oxygen; the consumption of LOX depends upon the activity of space programs. Pure oxygen is used in a multitude of breathing devices, submarines, diving bells, and in hospitals.

Chemical properties and reactions. The large values of the electronegativity and the electron affinity of oxygen are typical of elements that show only nonmetallic behaviour. In all of its compounds, oxygen assumes a negative oxidation state as is expected from the two half-filled outer orbitals. When these orbitals are filled by electron transfer, the oxide ion O^{2-} is created. In peroxides (species containing the ion O_2^{2-}) it is assumed that each oxygen has a charge of −1. This property of accepting electrons by complete or partial transfer defines an oxidizing agent. When such an agent reacts with an electron-donating substance, its own oxidation state is lowered. The change (lowering), from the zero to the −2 state in the case of oxygen, is called a reduction. Oxygen may be thought of as the "original" oxidizing agent, the nomenclature used to describe oxidation and reduction being based upon this behaviour typical of oxygen.

As described in the section on allotropy, oxygen forms the diatomic species, O_2, under normal conditions and, as well, the triatomic species ozone, O_3. There is some evidence for a very unstable tetratomic species, O_4. In the molecular diatomic form there are two unpaired electrons that lie in antibonding orbitals. The paramagnetic behaviour of oxygen confirms the presence of such electrons.

Reactivity of ozone

The intense reactivity of ozone is sometimes explained by suggesting that one of the three oxygen atoms is in an "atomic" state; on reacting, this atom is dissociated from the O_3 molecule, leaving molecular oxygen.

The molecular species, O_2, is not especially reactive at normal (ambient) temperatures and pressures. The atomic species, O, is far more reactive. The energy of dissociation ($O_2 \rightarrow 2O$) is large at 117.2 kilocalories per mole.

Sulfur. *Natural occurrence and distribution.* Many important metal ores are compounds of sulfur, either sulfides or sulfates. Some important examples are galena (lead sulfide, PbS), blende (zinc sulfide, ZnS), pyrite (iron disulfide, FeS_2), chalcopyrite (copper iron sulfide, $CuFeS_2$), gypsum (calcium sulfate dihydrate, $CaSO_4 \cdot 2H_2O$) and barite (barium sulfate, $BaSO_4$). The sulfide ores are valued chiefly for their metal content, although a process developed in the 18th century for making sulfuric acid utilized sulfur dioxide obtained by burning pyrite.

Allotropy. In sulfur, allotropy arises from two sources: (1) the different modes of bonding atoms into a single molecule and (2) packing of polyatomic sulfur molecules into different crystalline and amorphous forms. Some 30 allotropic forms of sulfur have been reported, but some of these probably represent mixtures. Only eight of the 30 seem to be unique; five contain rings of sulfur atoms and the others contain chains.

In the rhombohedral allotrope, designated ρ-sulfur, the molecules are composed of rings of six sulfur atoms. This form is prepared by treating sodium thiosulfate with cold, concentrated hydrochloric acid, extracting the residue with toluene, and evaporating the solution to give hexagonal crystals. ρ-sulfur is unstable, eventually reverting to orthorhombic sulfur (α-sulfur).

A second general allotropic class of sulfur is that of the eight-membered ring molecules, three crystalline forms of which have been well characterized. One is the orthorhombic (often improperly called rhombic) form, α-sulfur. It is stable at temperatures below 96° C. Another of the crystalline S_8 ring allotropes is the monoclinic or β-form, in which two of the axes of the crystal are perpendicular, but the third forms an oblique angle with the first two. There are still some uncertainties concerning its structure; this modification is stable from 96° C to the melting point, 118.9° C. A second monoclinic cyclooctasulfur allotrope is the γ-form, unstable at all temperatures, quickly transforming to α-sulfur.

An orthorhombic modification, S_{12} ring molecules, and still another unstable S_{10} ring allotrope are reported. The latter reverts to polymeric sulfur and S_8. At temperatures above 96° C, the α-allotrope changes into the β-allotrope. If enough time is allowed for this transition to occur completely, further heating causes melting to occur at 118.9° C; but if the α-form is heated so rapidly that the transformation to β-form does not have time to occur, the α-form melts at 112.8° C.

Effects of heating sulfur

Just above its melting point, sulfur is a yellow, transparent, mobile liquid. Upon further heating, the viscosity of the liquid decreases gradually to a minimum at about 157° C, but then rapidly increases, reaching a maximum value at about 187° C; between this temperature and the boiling point of 444.6° C, the viscosity decreases. The colour also changes, deepening from yellow through dark red, and, finally, to black at about 250° C. The variations in both colour and viscosity are considered to result from changes in the molecular structure. A decrease in viscosity as temperature increases is typical of liquids, but the increase in the viscosity of sulfur above 157° C probably is caused by rupturing of the eight-membered rings of sulfur atoms to form reactive S_8 units that join together in long chains containing many thousands of atoms. The liquid then assumes the high viscosity characteristic of such structures. At a sufficiently high temperature, all of the cyclic molecules are broken, and the length of the chains reaches a maximum. Beyond that temperature, the chains break down into small fragments. Upon vaporization, cyclic molecules (S_8 and S_6) are formed again; at about 900° C, S_2 is the predominant form; finally, monatomic sulfur is formed at temperatures above 1,800° C.

Commercial production. Elemental sulfur is found in volcanic regions as a deposit formed by the emission of hydrogen sulfide, followed by aerial oxidation to the element. Underground deposits of sulfur associated with salt domes in limestone rock provide a substantial portion of the world's supply of the element. These domes are located in the Louisiana swamplands of the United States and offshore in the Gulf of Mexico.

Frasch process

Herman Frasch, a German-born U.S. chemist, developed a process, subsequently known by his name, for extracting and raising pure sulfur from these deposits, which may be situated anywhere from a few hundred to several thousands of metres below the suface. Ordinary underground mining procedures are inapplicable since highly poisonous hydrogen sulfide gas accompanies the element in the domes.

The Frasch process takes advantage of the low melting point of sulfur, about 112° C. Water heated above this temperature (under pressure) is pumped down one of three concentric pipes, melting the sulfur. Compressed air is then forced down an inner pipe forming a froth of the molten sulfur, which is then forced up through the middle concentric pipe to the surface, where it is either pumped into bins for storage or into barges or ships for transporting to industrial areas for conversion, for the most part, to sulfuric acid. The availability of cheap, very pure sulfur from this process eliminated much of the need to mine sulfur from sulfides and volcanic sulfur-bearing rock for many years. By the mid-20th century the purification of sour (high sulfur-content) petroleum and improved methods for obtaining sulfur from metal sulfides had increased sulfur production from non-Frasch sources.

A few of the non-Frasch processes for sulfur production may be mentioned.

Other recovery processes

(1) Sulfur-bearing rock is piled into mounds. Shafts are bored vertically and fires set at the top of the shafts. The burning sulfur provides sufficient heat to melt the elemental sulfur in the rock layers below, and it flows out at the bottom of the pile. This is an old process, still used to some extent in Sicily. The product is of low purity and must be refined by distillation. The air pollution in the area of the process is so great that its operation is limited to certain times of the year when prevailing winds will carry the fumes away from populated areas.

(2) Rock bearing sulfur is treated with superheated water in retorts, melting the sulfur, which flows out. This process is a modification of the Frasch method.

(3) Sulfates (such as gypsum or barite) may be treated with carbon at high temperatures, forming the metal sulfides, CaS or BaS (the Chance–Claus process). The metal sulfides can be treated with acid, generating hydrogen sulfide, which in turn can be burned to give elemental sulfur.

(4) Tremendous tonnages of sulfur are available from smelter operations and from power production by combustion of fossil and sour petroleum fuels, some of which contain as much as 4 percent sulfur. Thus, generation of electrical power and heat represent a major source of atmospheric pollution by sulfur dioxide. Unfortunately, recovery and purification of sulfur dioxide from stack gases are expensive operations.

Wherever such metals as lead, zinc, copper, cadmium, or nickel (among others) are processed, much of the sulfuric acid needed in the metallurgical operations may be obtained on the site by converting sulfur dioxide, produced by roasting the ores, to sulfur trioxide, SO_3, and thence to sulfuric acid.

Sulfur available in bulk from commercial production usually is more than 99 percent pure, and some grades contain 99.9 percent sulfur. For research purposes, the proportion of impurities has been reduced to as little as one part in 10,000,000 by the application of procedures such as zone melting, column chromatography, electrolysis, or fractional distillation.

Conversion into sulfuric acid

Uses of sulfur. Sulfur is so widely used in industrial processes that its consumption often is regarded as a reliable indicator of industrial activity and the state of the national economy. Approximately six-sevenths of all the sulfur produced is converted into sulfuric acid, for which the largest single use is in the manufacture of fertilizers (phosphates and ammonium sulfate). Other important uses include the production of pigments, detergents, fibres, petroleum products, sheet metal, explosives, and storage batteries; hundreds of other applications are known. Sulfur not converted to sulfuric acid is used in making paper, insecticides, fungicides, dyestuffs, carbon disulfide (a solvent employed in making rayon, cellophane, and industrial chemicals), and numerous other products.

Selenium. *Occurrence and distribution.* The proportion of selenium in the Earth's crust is about 10^{-5} to 10^{-6} percent. It has been obtained mainly from the anode slimes (deposits and residual materials from the anode) in electrolytic refining of copper and nickel. Other sources are the flue dusts in copper and lead production and the gases formed in roasting pyrites. Selenium accompanies copper in the refining of that metal: about 40 percent of the selenium present in the original ore may concentrate in copper deposited in electrolytic processes. About 1.5 kilograms of selenium can be obtained from a ton of smelted copper.

Allotropy. The allotropy of selenium is not as extensive as that of sulfur, and the allotropes have not been studied as thoroughly. Only two crystalline varieties of selenium are composed of cyclic Se_8 molecules: designated α and β, both exist as red monoclinic crystals. A gray allotrope having metallic properties is formed by keeping any of the other forms at 200°–220° C.

An amorphous (noncrystalline), red, powdery form of selenium results when a solution of selenious acid or one of its salts is treated with sulfur dioxide. If the solutions are very dilute, extremely fine particles of this variety yield a transparent red colloidal suspension. Clear red glass results from a similar process that occurs when molten glass containing selenites is treated with carbon. A glassy, almost black variety of selenium is formed by rapid cooling of other modifications from temperatures above 200° C. Conversion of this vitreous form to the red, crystalline allotropes takes place upon heating it above 90° C or upon keeping it in contact with organic solvents, such as chloroform, ethanol, or benzene.

Preparation. Pure selenium is obtained from the slimes and sludges formed in producing sulfuric acid. The impure red selenium is dissolved in sulfuric acid in the presence of an oxidizing agent, such as potassium nitrate or certain manganese compounds. Both selenious acid, H_2SeO_3, and selenic acid, H_2SeO_4, are formed and can be leached from residual insoluble material. Other methods utilize oxidation by air (roasting) and heating with sodium carbonate to give soluble sodium selenite, $Na_2SeO_3 \cdot 5H_2O$, and sodium selenate, Na_2SeO_4. Chlorine may also be employed: its action upon metal selenides produces volatile compounds including selenium dichloride, $SeCl_2$; selenium tetrachloride, $SeCl_4$; diselenium dichloride, Se_2Cl_2; and selenium oxychloride, $SeOCl_2$. In one process, these selenium compounds are converted by water to selenious acid. The selenium is finally recovered by treating the selenious acid with sulfur dioxide.

Selenium is a common component of ores valued for their content of silver or copper; it becomes concentrated in the slimes deposited during electrolytic purification of the metals. Methods have been developed to separate selenium from these slimes, which also contain some silver and copper. Melting the slime forms silver selenide, Ag_2Se, and copper(I) selenide, Cu_2Se. Treatment of these selenides with hypochlorous acid, HOCl, gives soluble selenites and selenates, which can be reduced with sulfur dioxide. Final purification of selenium is accomplished by repeated distillation.

Photoconductivity of selenium

Physical–electrical properties. The most outstanding physical property of crystalline selenium is its photoconductivity: on illumination, the electrical conductivity increases more than 1,000-fold. This phenomenon results from the promotion or excitation of relatively loosely held electrons by light to higher energy states (called the conduction levels), permitting electron migration and, thus, electrical conductivity. In contrast the electrons of typical metals are already in conduction levels or bands, able to flow under the influence of an electromotive force.

The electrical resistivity of selenium varies over a tremendous range, depending upon such variables as the nature of the allotrope, impurities, the method of refining, temperature, and pressure. Most metals are insoluble in selenium, and nonmetallic impurities increase the resistivity.

Illumination of crystalline selenium for 0.001 second increases its conductivity by a factor of 10 to 15 times. Red light is more effective than light of shorter wave length.

Advantage is taken of these photoelectric and photosensitivity properties of selenium in the construction of a variety of devices that can translate variations in light intensity into electric current and thence to visual, magnetic, or mechanical effects. Alarm devices, mechanical opening and closing devices, safety systems, television, sound films, and xerography depend upon the semiconducting property and photosensitivity of selenium. Rectification of alternating electrical current (conversion into direct current) has for years been accomplished by selenium-controlled devices. Many photocell applications using selenium have been replaced by other devices using materials more sensitive, more readily available, and more easily fabricated than selenium.

Tellurium. *Occurrence and preparation.* The demand for tellurium does not match that for selenium. The two elements are found together in many ores; they may be isolated by employing the processes described in connection with selenium, obtaining solutions containing salts of both selenious and tellurous acids, H_2SeO_3 and H_2TeO_3. Upon treatment of these solutions with sulfuric acid, tellurium dioxide, TeO_2, separates because of its low solubility, while the selenious acid remains dissolved. The tellurium dioxide can be converted into elemental tellurium by treatment with sulfur dioxide; an electrolytic process is used to purify the product.

Physical and chemical properties. In tellurium, the covalent bonding necessary to provide large ring- and chain-molecules by catenation is almost nonexistent. The element crystallizes in the rhombohedral form. It is silvery white and isomorphous with gray selenium—that is, the structure and dimensions of the crystals are very similar. It is brittle but not very hard. The tellurium atoms form spiral chains in the crystal with Te–Te distances of 3.74 Å. There are no good solvents for tellurium, although certain compounds oxidize or reduce the element to soluble substances. The photoconductivity of tellurium is not as pronounced as that of selenium, and tellurium does not have major industrial uses.

Polonium. The existence of polonium in pitchblende, an ore of uranium, was noted by the Curies. Polonium is extremely rare, even in pitchblende: 1,000 tons of the ore must be processed to obtain 40 milligrams of polonium. In the chemical isolation, the ore is treated with hydrochloric acid, and the resulting solution is heated with hydrogen sulfide to precipitate polonium monosulfide, PoS, along with other metal sulfides, such as that of bismuth, Bi_2S_3, which resembles polonium monosulfide closely in chemical behaviour, though it is less soluble. Because of the difference in solubility, repeated partial precipitation of the mixture of sulfides concentrates the polonium in the more soluble fraction, while the bismuth accumulates in the less soluble portions. The difference in solubility is small, however, and the process must be repeated many times to achieve a complete separation. Purification is accomplished by electrolytic deposition.

Purification process

Two modifications of polonium are known, an α- and a β-form, both of which are stable at room temperature and possess metallic character. The fact that its electrical conductivity decreases as the temperature increases places polonium among the metals rather than the metalloids or nonmetals. (R.C.Br./Ed.)

Halogen elements

Fluorine (F), chlorine (Cl), bromine (Br), iodine (I), and astatine (At) are the members of the family of halogen elements, and they constitute Group VIIa of the periodic table (see Figure 10). They were given the name halogen from the Greek roots *hal-* ("salt") and *gen* ("to produce"), because they all produce sodium salts of similar properties, of which sodium chloride, table salt, is the best known. Because of their great reactivity, the free halogen elements are not found in nature. In combined form, fluorine is the most abundant of the halogens in the Earth's crust. The percentages of the halogens in the igneous rocks of the Earth's crust are 0.06 fluorine, 0.031 chlorine, 0.00016 bromine, and 0.00003 iodine. Astatine does not occur in nature because it consists only of short-lived radioactive isotopes.

The halogen elements show great resemblances to one another in their general chemical behaviour and in the properties of their compounds with other elements. There is, however, a progressive change in properties from fluorine through chlorine, bromine, and iodine to astatine—the difference between two successive elements being most pronounced with fluorine and chlorine. Fluorine is the most reactive of the halogens and, in fact, of all elements, and it has certain other properties that set it apart (see below *General properties of the halogens*).

Chlorine is the best known of the halogen elements. The free element is widely used as a water purification agent, and it is employed in a number of chemical processes. Sodium chloride, of course, is one of the most familiar of chemical compounds. Fluorides are known chiefly for the controversy over their addition to public water supplies to prevent tooth decay, but organic fluorides are also used as refrigerants and lubricants. Iodine is most familiar as an antiseptic, and bromine is used chiefly to prepare a gasoline additive (ethylene dibromide) that prevents deposits of lead in engines.

HISTORY

Rock salt (common salt, sodium chloride) has been known for several thousand years; it is the main constituent of the salts dissolved in seawater, from which it was obtained in ancient Egypt by evaporation. In 1648 the German chemist Johann Rudolf Glauber obtained a strong acid, which he called spirit of salt, by heating moist salt in a charcoal furnace and condensing the fumes in a receiver. Later he obtained the same product, now known to be hydrochloric acid, by heating salt with sulfuric acid.

In 1774 the Swedish chemist Carl Wilhelm Scheele treated powdered black oxide of manganese with hydrochloric acid and obtained a greenish-yellowish gas, which he failed to recognize as an element. The true nature of the gas as an element was recognized in 1810 by Sir Humphry Davy, who later named it chlorine and provided an explanation for its bleaching action.

In 1811 the French chemist Bernard Courtois obtained a violet vapour by heating seaweed ashes with sulfuric acid. This vapour condensed to a black crystalline substance, which he called "substance X." In 1813 Davy, who was passing through Paris on his way to Italy, recognized substance X as an element analogous to chlorine; he suggested the name iodine.

Discovery of bromine

Bromine was discovered in 1826 by the French chemist Antoine-Jérôme Balard in the residues from the manufacture of sea salt at Montpellier. He liberated the element by passage of chlorine through aqueous solution of the residues, which contained magnesium bromide. Distillation of the material with manganese dioxide and sulfuric acid produced red vapours, which were condensed to a dark liquid. The similarity of this procedure to that for making chlorine suggested to Balard that he had obtained a new element similar to chlorine. (The German chemist Justus von Liebig appears to have obtained the element before Balard, but he wrongly considered it to be iodine chloride.)

The fluorine-containing mineral fluorspar (fluorite) was described in 1529 by Georg Bauer, the German physician and mineralogist better known as Agricola. It appears likely that crude hydrofluoric acid was first prepared by an unknown English glassworker in 1720. In 1771 Scheele obtained hydrofluoric acid in an impure state by heating fluorspar with concentrated sulfuric acid in a glass retort, which was greatly corroded by the product; as a result, vessels made of metal were used in subsequent experiments with the substance. The nearly anhydrous acid was prepared in 1809, and two years later the French physicist André-Marie Ampère suggested that it was a compound of hydrogen with an unknown element, analogous to chlorine, for which he suggested the name fluorine. Fluorspar was then recognized to be calcium fluoride.

Isolation of fluorine

The isolation of fluorine was for a long time one of the chief unsolved problems in inorganic chemistry, and it was not until the year 1886 that the French chemist Henri Moissan prepared the element by electrolyzing a solution of potassium hydrogen fluoride in hydrogen fluoride. The difficulty in handling the element and its toxic properties contributed to the slow progress in fluorine chemistry. Indeed, up to the time of World War II the element appeared to be a laboratory curiosity. Then, however, the use of uranium(VI) fluoride in the separation of uranium isotopes and the development of organic fluorine compounds of industrial importance made fluorine an industrial chemical of considerable use. Astatine was prepared for the first time in 1940 by bombarding bismuth with alpha particles.

GENERAL PROPERTIES OF THE GROUP

Oxidizing property. Probably the most important generalization that can be made about the halogen elements is that they are all oxidizing agents; *i.e.,* they raise the oxidation state, or oxidation number, of other elements—a property that used to be equated with combination with oxygen but that is now interpreted in terms of transfer of electrons from one atom to another. In oxidizing another element, a halogen is itself reduced—*i.e.,* its oxidation number is reduced from 0, the typical state for a free element, to -1. In this state the halogens combine with other elements to form compounds known as halides—namely, fluorides, chlorides, bromides, iodides, and astatides. Many of the halides may be considered to be salts of the respective hydrogen halides, which are colourless

Formation of saltlike compounds

gases at room temperature and atmospheric pressure and which are known to exhibit acidic properties. Indeed the general term salt is derived from rock salt, or table salt (sodium chloride). The tendency of the halogen elements to form saltlike compounds increases in the following order: astatine, iodine, bromine, chlorine, and fluorine.

This order is also that of increasing oxidizing properties of the free halogen elements. Fluorides are usually more stable than the corresponding chlorides, bromides, or iodides. (Often astatine is omitted from general discussions of the halogens because it is less well known than the other elements.) Conversely, of the halogen elements, fluorine is prepared in the free state with the greatest difficulty and iodine with the least. As a class, the halogen elements are nonmetals, but astatine shows certain properties resembling those of the metals.

Electronic structure. The chemical behaviour of the halogen elements can be discussed most conveniently in terms of their position in the periodic table of the elements. In the periodic table the halogens comprise Group VIIa, the group immediately preceding the noble gases. It is characteristic of the atoms of the halogen elements that each carries seven electrons in its outermost orbitals. The seven outermost electons of an atom of a halogen element are in two different kinds of orbitals, designated *s* (with two electrons) and *p* (with five). Potentially, a halogen atom could hold one more electron (in a *p* orbital), which would give the electrons the same arrangement (configuration) as that of the noble gas next to it in the periodic table, the noble gases all having exceptionally stable configurations of electrons in their atoms. It is in acquiring such an electron that an atom of a halogen element acts as an oxidizing agent (or is itself reduced). In the process, the atom acquires an electrical charge (in this case a negative one) and becomes a charged particle or ion. Halogen elements exist in their salts as halide ions, which are relatively large ions, colourless and extremely stable.

Diatomic halogen molecules

At room temperature and atmospheric pressure the halogen elements in their free states exist as diatomic molecules—molecules composed of two atoms each. In molecular fluorine the atoms are held together by a bond made from the union of a *p* orbital from each atom, such a bond being classed as a sigma bond. Beginning with chlorine, however, the halogen atoms are able to make use of different orbitals, called *d* orbitals—an option that is not open to the fluorine atoms. By forming partial bonds with their *d* orbitals, chlorine and the other halogens (with the exception of fluorine) form diatomic molecules held in part by multiple bonds (called pi bonds), which are stronger than single bonds, such as those that hold together the atoms in a fluorine molecule.

Relative reactivity. The great reactivity of fluorine largely stems from the relatively low energy of the fluorine–fluorine bond (37.7 kilocalories per mole, a standard measure for bond energies). The chlorine–chlorine bond is considerably stronger (57 kilocalories per mole) because of its partial multiple-bond character.

Fluorine and chlorine are gases at room temperature. Bromine is a reddish-brown liquid at room temperature, and is—apart from mercury—the only element that is liquid at 20° C and atmospheric pressure. Iodine forms dark-violet crystals under these conditions. In the solid state the halogen elements form molecular lattices—extended weblike structures including a great many separate units. The lattice energies increase with increasing size of the molecules. The principal properties of the halogen elements are noted in Table 21.

Electron affinities

The energy released in the formation of an ion from a free atom and an electron (brought up from an infinite distance) is called the electron affinity. The electron affinities for the halogen elements are all high and show only slight differences from one another. It is known, however,

Table 21: Properties of the Halogen Elements

	fluorine	chlorine	bromine	iodine	astatine
Atomic number	9	17	35	53	85
Atomic weight	18.9984	35.453*	79.904*	126.904	(210)
Colour of element	light greenish yellow	greenish yellow	brown-red	dark violet	—
Melting point (°C)	−219.62	−100.98	−7.25	113.5	~ 300
Boiling point (°C)	−188.14	−34.05	58.8	184.4	~ 370
Density (760 mm Hg)					
Gas (g/l)	1.696 (0° C)	3.214 (0° C)	5.6 (175.5° C)	6.75 (185° C)	—
Liquid (g/cm^3)	1.108 (−188° C)	1.655 (−70° C)	3.12 (20° C)	3.960 (120° C)	—
Solid (g/cm^3)	1.32 (0° K)	2.17 (−195° C)	4.17 (0° K)	4.930 (20° C)	—
Solubility in water	(reacts)	2.3 (vol/vol 20° C) carbon tetrachloride (reacts) 8.48% (by weight, 20° C) chloroform (280 g/1,000g, 0° C)	3.41 (g/100 g, 20°) completely miscible	0.293 (g/1,000 g aq. iodide solution completely miscible 19.2 (g/1,000g) completely miscible 49.7 (g/1,000g)	—
Oxidation numbers	−1	−1, +1, +3, +4, +5, +6, +7	−1, +1, +3, +4, +5, +7	−1, +1, +3, +4, +5, +6, +7	−1, +1, +3, +5, +7
Electronic configuration	(He) $2s^22p^5$	(Ne) $3s^23p^5$	(Ar) $3d^{10}4s^24p^5$	(Kr) $4d^{10}5s^25p^5$	(Xe)$4f^{14}5d^{10}6s^26p^5$
Isotopic abundance (terrestrial percent)	^{19}F (100)	^{35}Cl (75.53) ^{37}Cl (24.47)	^{79}Br (50.54) ^{81}Br (49.46)	^{127}I (100)	—
Radioactive isotopes (mass numbers)	17, 18, 20–22	32–34, 36, 38–40	74–78, 80, 82-90	117–126, 128–139	200–219
Heat of fusion (kcal/mol)	0.244	1.53	2.58	3.65	—
Heat of vaporization (kcal/mol)	1.561	4.863	7.159	9.889	—
Heat of hydration of X^- (kcal/mol)	120.8	88	80.3	70.5	—
Specific heat (kcal/g/°C at 25°C)					
Constant pressure	0.1842 (17° C)	0.114	0.0537	0.034	—
Constant volume		0.0849	0.04257	0.02697	—
Critical temperature (°C)	−129	144	302	512	—
Critical pressure (atm)	56.8	77.3	126	116†	—
Critical density (g/cm^3)	0.63	0.567	1.064	1.336	—
Crystal structure	—	orthorhombic	orthorhombic	orthorhombic	—
Electrical resistivity (microohms-cm)	—	~ 10^{10} (near freezing point)	6.5×10^{10} (25° C)	5.85 (25° C)	—
Magnetic susceptibility (cm^3/g)					
Gas	-3.4×10^{-7}	-18.7×10^{-10} (0° C, 760 mm Hg)	-73.5×10^{-6}	-3.9×10^{-7} (118° C)	—
Liquid		-5.9×10^{-7} (−16° C)	-56.4×10^{-6}	-88.7×10^{-6} (sol., 28° C)	—
Radius					
Ionic (Å)	1.33	1.81	1.96	2.20	~ 2.27
Covalent (Å)	0.71	0.99	1.14	1.33	~ 1.4
Bond energy (kcal/mol)	37.7	58.0	46.1	36.1	—
Ionization energy (kcal/mol)	402	300	273	241	—
Electron affinity (kcal/mol)	79.5	83.3	77.5	70.6	—
Electronegativity (Allred-Rochow)	4.10	2.83	2.74	2.21	1.96

*Variation of isotopic abundance in terrestrial samples limits the precision of the atomic weight given. †Estimated.

that the oxidizing properties (ability to take up an electron) increase from astatine to fluorine, and this increase is reflected in the position of the respective oxidation–reduction systems in the electromotive series (a standard arrangement of oxidation–reduction systems in water in order of oxidizing strength).

The fluoride ion is extremely stable toward oxidation; the iodide ion is a mild reducing agent. This difference between the ions is caused by two factors: (1) a decreasing tendency for the halide ion to act as an electron-pair donor (a so-called Lewis base) in the series fluoride, chloride, bromide, iodide, and astatide, leading to a decrease in stability of the hydrated form of the ion (*i.e.,* the free ion surrounded by water molecules) in the same order; and (2) a decrease in electron pair acceptor (Lewis acid) properties of the free halogens in the reverse order, namely, astatine, iodine, bromine, chlorine, and fluorine.

Within a molecule in which atoms are held together by a shared electron pair—*i.e.,* by a covalent or nonionic bond—the tendency of an atom to attract the shared electrons may be expressed by an electronegativity value. Fluorine has the greatest electronegativity value (is the most electronegative) of all elements (not just the halogens), and there is a decrease in electronegativity within the family of the halogen elements from fluorine through chlorine, bromine, and iodine to astatine.

Fluorine replaces any other halide ion from its compounds, as shown in the following equations (in which atoms of the elements are represented by their chemical symbols; number of atoms of a given kind in a molecule are indicated by subscripts; and positive and negative charges on ions are indicated by superscript plus and minus signs; as is customary in chemical equations the reactants are written to the left and the products to the right; the double arrows indicate that both the forward reaction and the reverse reaction occur). Chlorine, however, replaces only bromide, iodide, and astatide ions, and bromine only iodide and astatide ions. Free fluorine, chlorine, bromine, and iodine are expected to replace astatide ions.

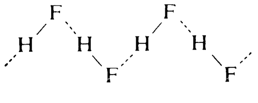
$$F_2 + 2Cl^- \rightleftharpoons 2F^- + Cl_2$$

$$F_2 + 2Br^- \rightleftharpoons 2F^- + Br_2$$

$$F_2 + 2I^- \rightleftharpoons 2F^- + I_2$$

$$F_2 + 2At^- \rightleftharpoons 2F^- + At_2.$$

The halogen elements all form compounds with hydrogen—the hydrogen halides. The energy of the hydrogen–halogen bond increases strongly from iodide to fluoride. Hydrogen fluoride in the crystalline state consists of infinite zigzag chains, as shown in the diagram



in which H represents the hydrogen atoms and (as before) F the fluorine atoms; the solid lines represent covalent bonds between the hydrogen and fluorine atoms within the molecules, and the dotted lines represent secondary bonds, called hydrogen bonds. The hydrogen bonds between hydrogen fluoride molecules are considerably weaker (seven kilocalories per mole) than those within the molecules (135 kilocalories per mole), yet they are retained to a great extent in the liquid state. Similar hydrogen bonding exists in the other hydrogen halides, but it is considerably weaker. The large difference in hydrogen bonding between hydrogen fluoride and the other hydrogen halides accounts for the relatively high melting and boiling points of hydrogen fluoride as compared to those of hydrogen chloride and the other hydrogen halides. The hydrogen–halogen bond energies also decrease considerably in going from hydrogen fluoride to hydrogen iodide.

Ionization potentials of halogens

The ionization potentials of the halogens are generally high, but they fall markedly with atomic number. Fluorine is the only halogen that does not form compounds with positive oxidation states—*i.e.,* states in which it has lost, rather than gained, electrons. This property is related to the inability of the fluorine atom to use *d* orbitals in bonding (see above), an inability not shared by the other halogen elements. Iodine forms an appreciable number of compounds in which a unipositive iodine ion is stabilized by coordination (surrounding with other groups); for example, with the coal-tar base, pyridine. The same situation has also been found with astatine.

Whereas fluorine exhibits only the single oxidation state of −1, the principal oxidation states of chlorine, bromine, and iodine are −1, +1, +3, +5, and +7. The oxo acids are compounds in which halogen atoms are joined to oxygen atoms and show large positive oxidation numbers. The oxo acids are all powerful oxidizing agents, being reduced to the corresponding aqueous hydrogen halides—the oxidation numbers passing from +5 or +7 to −1 in the process. The oxidizing properties of the oxo acids, however, fall with increasing oxidation number of the halogen because the coordination number increases with increasing oxidation number and the ion is stabilized by the coordination.

All the molecules and ions in which halogen atoms employ four electron pairs in bonding are basically tetrahedral in shape—as is, for example, the perchlorate ion $(ClO_4)^-$. Those employing five electron pairs are square pyramidal in structure, such as iodine(V) fluoride (IF_5)—the roman numeral in the name being used to indicate the oxidation number of the element it follows. Those with six electron pairs are octahedral; *e.g.,* the paraperiodate ion, $(IO_6)^{5-}$. The unique binary compound iodine(VII) fluoride (IF_7) is believed to have a pentagonal bipyramidal arrangement of fluorine atoms.

The maximum coordination number of chlorine(VII) toward oxygen is 4 (*i.e.,* the chlorine atom is surrounded by four oxygen atoms) in the perchlorate ion, $(ClO_4)^-$, whereas that of iodine(VII) is 6 in the paraperiodate ion, $(IO_6)^{5-}$. Analagous behaviour is found in the elements preceding the respective halogen in the periodic table, sulfur and tellurium.

INDIVIDUAL HALOGEN GROUP ELEMENTS

Fluorspar—source of fluorine

Fluorine. *Occurrence and distribution.* The fluorine-containing mineral fluorspar has been used for centuries as a flux (cleansing agent) in various metallurgical processes. The name fluorspar is derived from the Latin *fluere,* "to flow." The mineral subsequently proved to be a source of the element, which was named fluorine accordingly. The colourless, transparent crystals of fluorspar exhibit a bluish tinge when illuminated, and this property is accordingly known as fluorescence.

Fluorine is found in nature only in the form of its chemical compounds, except for trace amounts of the free element in fluorspar that has been subjected to the action of radiation from radium. The principal fluorine-containing minerals are: (1) fluorspar (fluorite, CaF_2), deposits of which occur in Illinois, Kentucky, Derbyshire, southern Germany, the south of France, and the Soviet Union; (2) cryolite (Na_3AlF_6), chiefly from Greenland; (3) fluoroapatite $(Ca_5[PO_4]_3[F,Cl])$, widely distributed and containing variable amounts of fluorine and chlorine; (4) topaz $Al_2SiO_4[F,OH]_2)$, the gemstone; and (5) lepidolite, a mica as well as a component of animal bones and teeth.

Production and use. Fluorspar is the most important source of fluorine. In the manufacture of hydrogen fluoride, powdered fluorspar is distilled with concentrated sulfuric acid in a lead or cast-iron apparatus. The hydrogen fluoride is obtained in a fairly anhydrous state by fractional distillation in copper or steel vessels and stored in steel cylinders. The usual impurities in commercial hydrogen fluoride are sulfurous and sulfuric acids as well as fluorosilicic acid, arising from the presence of silica in the fluorspar. Traces of moisture may be removed by electrolysis with platinum electrodes.

The preparation of the free element is carried out by electrolytic procedures in the absence of water. Generally these take the form of electrolysis of a melt of potassium fluoride–hydrogen fluoride (in a ratio of 1 to 2.5–5) at temperatures between 30° and 70° C, 80° and 120° C, or at 250° C. During the process the hydrogen fluoride content of the electrolyte is decreased and the melting point rises; it is therefore necessary to add hydrogen fluoride. In the high-temperature cell the electrolyte is replaced when the melting point reaches over 300° C. Fluorine can be safely stored under pressure in cylinders of stainless steel if the valves of the cylinders are free from traces of organic matter.

Uses in fluorination

The element is used for the preparation of various fluorides, such as chlorine trifluoride, sulfur(VI) fluoride, or cobalt(III) fluoride. The cobalt compound and certain other metal fluorides are important fluorinating agents for organic compounds. With appropriate precautions, the element itself may be used for the fluorination of organic compounds. Fluorine derivatives of hydrocarbons (compounds of carbon and hydrogen) are useful as refrigerants and as lubricants. The element is also used for the preparation of uranium(VI) fluoride, which is important in the process of separation of uranium-235 from natural uranium. Its outstanding oxidizing properties make elemental fluorine of possible interest as an oxidizer in rocket propulsion. Hydrogen fluoride and boron(III) fluoride are produced commercially because they are good catalysts for the alkylation reactions used to prepare organic compounds of many kinds. The addition of sodium fluoride to drinking water to reduce the incidence of dental caries in children is a matter of some controversy.

Chemical properties. At room temperature fluorine is a pale-yellow gas with an irritating odour. Inhalation of the gas is dangerous except in very low concentration. There is only one stable isotope of the element—fluorine-19.

Because fluorine is the most electronegative of the elements, atomic groupings rich in fluorine are negatively charged. Methyl iodide (CH_3I) and trifluoroiodomethane (CF_3I) have different properties for that reason. The charge distributions in these molecules are shown in the following formulas, in which the Greek symbol δ (delta) indicates a partial charge:

$$\overset{\delta+}{CH_3}\overset{\delta-}{I} \qquad\qquad \overset{\delta-}{CF_3}\overset{\delta+}{I}.$$

The ionization potential of fluorine is very high (402 kilocalories per gram atom), giving a standard heat formation for fluoride ion of 420 kilocalories per mole.

The small size of the fluorine atom makes it possible to pack a relatively large number of fluorine atoms or fluoride ions around a given coordination centre—*i.e.,* around a central atom. Fluorine is the most powerfully oxidizing element. No other substance, therefore, is able to reduce the fluoride ion to the free element, and for this reason the element is not found in the free state in nature. Furthermore, all chemical methods have failed to produce the element, success having been achieved only by the use of electrolytic methods. The high oxidizing properties of fluorine allow the element to produce the highest oxidation numbers possible in other elements, and many higher fluorides of elements are known for which there are no other corresponding halides—*e.g.,* silver(II) fluoride, cobalt(III) fluoride, rhenium(VII) fluoride, bromine(V) fluoride, and iodine(VII) fluoride.

Fluorine reacts with nearly all other elements at room temperature. Some metals, such as nickel, are quickly covered by a fluoride layer, which prevents further attack of the metal by the element. When dry, certain metals, such as mild steel, copper, aluminum, or Monel metal, are not attacked by fluorine to any extent at ordinary temperatures. For work with fluorine at temperatures up to 600° C Monel metal is suitable; sintered alumina is resistant up to 700° C. When lubricants are required, fluorocarbon oils are most suitable.

Fluorine reacts violently with organic matter (such as rubber, wood, and cloth), and controlled fluorination of organic compounds by the action of elemental fluorine is not possible unless special precautions are taken.

Analysis. The accurate quantitative determination of the amount of fluorine in compounds is difficult. Free fluorine may be detected by its oxidizing action on chlorides such as sodium chloride, as shown:

$$F_2 + 2NaCl \rightleftharpoons 2NaF + Cl_2.$$

The chlorine evolved may be determined by measurement of the volume of gas produced. The principal qualitative tests for the presence of fluorine are: (1) liberation of hydrogen fluoride by the action of sulfuric acid, (2) formation of a precipitate of calcium fluoride upon addition of a calcium chloride solution, and (3) decoloration of a yellow solution prepared from titanium(IV) oxide and hydrogen peroxide in sulfuric acid. Quantitative methods for analyzing fluorine are: (1) precipitation of calcium fluoride in the presence of sodium carbonate and treatment of the precipitate with acetic acid; (2) precipitation of lead chlorofluoride by addition of sodium chloride and lead nitrate; and (3) titration (determination of concentration of a dissolved substance) with thorium nitrate [$Th(NO_3)_4$] solution using sodium alizarin sulfonate as indicator, according to the equation.

Qualitative tests for fluorine

$$Th(NO_3)_4 + 4KF \rightleftharpoons ThF_4 + 4KNO_3.$$

Chlorine. *Occurrence and distribution.* When Sir Humphry Davy in 1810 recognized the elementary nature of the yellowish-green gas that had first been obtained by Scheele in 1774, he suggested the name chlorine from the Greek *chloros,* meaning "yellowish-green." Apart from very small amounts of free chlorine in volcanic gases, chlorine is found only in the form of chemical compounds. It constitutes 0.031 percent of the Earth's crust. The most common compound of chlorine is sodium chloride, which is found in nature as crystalline rock salt, often discoloured by impurities. Sodium chloride is also present in seawater, which has an average concentration of about 3 percent of that salt. Certain landlocked seas, such as the Dead Sea, contain up to 7.2 percent of dissolved salt. Besides sodium chloride, other metal halides, such as magnesium chloride, magnesium bromide, and—in small amount—certain sulfates, are contained in seawater. Small quantities of sodium chloride are present in blood and in milk. Other chlorine-containing minerals are sylvite (potassium chloride, KCl), bischofite ($MgCl_2 \cdot 6H_2O$), carnallite ($KCl \cdot MgCl_2 \cdot 6H_2O$), and kainite ($KCl \cdot MgSO_4 \cdot 3H_2O$). Free hydrochloric acid is present in the stomach.

Present-day salt deposits must have been formed by evaporation of prehistoric seas, the salts with the least solubility in water crystallizing first, followed by those with greater solubility. Because potassium chloride is more soluble in water than sodium chloride, certain rock salt deposits—such as those at Stassfurt, East Germany—were covered by a layer of potassium chloride. In order to gain access to the sodium chloride, the potassium salt, important as a fertilizer, is removed first.

Mining of rock salt deposits

Production and use. Rock salt deposits are usually mined; occasionally water is pumped down and brine, containing about 25 percent sodium chloride, is brought to the surface. When the brine is evaporated, impurities separate first and can be removed. In warm climates salt is obtained by evaporation of shallow seawater by the sun, to give bay salt.

Chlorine is produced on a large scale by any of a number of different methods:

1. By electrolysis of a concentrated solution of sodium chloride in water. Hydrogen is evolved at the cathode and chlorine at the anode. At the same time, an alkali metal hydroxide is produced in the electrolyte, and hence this process is often referred to as chlorine-alkali-electrolysis.

The chemical reactions occurring at each electrode and the overall cell process are given in the equations below:

cathode process: $2H_2O + 2e^- \longrightarrow 2OH^- + H_2$
(cathode of iron)
anode process: $2Cl^- \longrightarrow Cl_2 + 2e^-$
(anode of graphite)
cell process: $2H_2O + 2Cl^- \longrightarrow 2OH^- + H_2 + Cl_2$

in which the symbol e^- represents a single electron. In the reaction vessel free chlorine and hydroxide ions must not come in contact with each other because chlorine would be consumed according to the reaction:

$$Cl_2 + 2OH^- \rightarrow (ClO)^- + Cl^- + H_2O.$$

To accomplish the separation of chlorine gas and the hydroxide ion, a porous wall is inserted between the electrodes (diaphragm process), or the iron cathode is replaced by a cathode consisting of liquid mercury (mercury cathode process) which avoids the production of hydroxide ions at the electrode. Instead, free sodium is discharged at the cathode, and this metal is readily dissolved in the mercury forming an amalgam, as follows:

$$2Na^+ + 2e^- \rightleftharpoons 2Na(\text{amalgam}).$$

The amalgam is allowed to react with water outside the cell:

$$2Na(\text{amalgam}) + 2H_2O \longrightarrow 2Na^+ + 2OH^- + H_2.$$

The overall process is equivalent to the cell process given above.

2. By electrolysis of fused sodium chloride, which also produces metallic sodium; chlorine is again evolved at the anode.

3. By electrolysis of fused magnesium chloride, in which chlorine is formed as a by-product in the manufacture of metallic magnesium.

4. By oxidation of hydrogen chloride. In this process gaseous hydrogen chloride mixed with air or oxygen is passed over pumice in contact with cupric chloride as a catalyst, as shown in the following equation:

$$4HCl + O_2 \xrightarrow{\text{catalyst}} 2H_2O + 2Cl_2.$$

The equilibrium constant for this reaction decreases with increase of temperature; *i.e.,* the reaction proceeds less extensively at higher temperatures. In practice, however, a temperature of 400° C is required to achieve a reasonable rate of conversion.

5. Of historical interest is the process in which a mixture of almost any solid chloride and manganese dioxide (MnO_2) yields chlorine when heated with concentrated sulfuric acid (H_2SO_4). The reaction occurs, as follows:

$$2NaCl + 3H_2SO_4 + MnO_2 \rightarrow MnSO_4 + 2NaHSO_4 + 2H_2O + Cl_2.$$

Laboratory preparations of chlorine

In the laboratory chlorine is frequently prepared by the oxidation of concentrated hydrochloric acid with permanganate or dichromate salts:

$$2MnO_4^- + 10Cl^- + 16H^+ \rightarrow 2Mn^{2+} + 5Cl_2 + 8H_2O$$

$$Cr_2O_7^{2-} + 6Cl^- + 14H^+ \rightarrow 2Cr^{3+} + 3Cl_2 + 7H_2O.$$

Similarly, chlorine may be prepared on a small scale by addition of any dilute acid to bleaching powder, a mixture of dry calcium salts, including the hydroxide, chloride, and hypochlorite.

Most of the chlorine produced is used for chemical processes involving the introduction of chlorine into organic compounds, yielding carbon tetrachloride (used as a solvent, a fire extinguisher, and a dry-cleaning agent) and glycols (used as antifreeze), and other organic compounds for the manufacture of plastics (polyvinyl chloride) and synthetic rubber. Sulfur chloride, made by the action of chlorine on carbon disulfide or by combining sulfur and chlorine, is used in the vulcanization of rubber and as a chlorinating agent in organic synthesis. Chlorine and carbon monoxide form carbonyl chloride, or phosgene, which has been employed as a war gas and is used in metallurgy to transform certain oxides into chlorides. Much chlorine is used to sterilize water and wastes; and the substance is employed either directly or indirectly as a bleaching agent for paper or textiles. Chlorine is applied in the manufacture of hydrochloric acid, the extraction of titanium with formation of titanium(IV) chloride, the removal of tin from old tinplate, and in the production of antiknock materials for gasolines, such as tetraethyl lead. Anhydrous aluminum chloride is made by the reaction of chlorine with scrap aluminum or with aluminum oxide and carbon. Chlorine is also used to prepare silicon(IV) chloride and methyl chloride, which are employed in the synthesis of silicon materials. Chlorine enters directly, or indirectly as an intermediate, into many organic syntheses of industrial importance.

Physical and chemical properties. Chlorine is a greenish-yellow gas at room temperature and atmospheric pressure. It is considerably heavier than air. It has a choking smell, and inhalation causes suffocation, constriction of the chest, tightness in the throat, and—after severe exposure—edema (filling with fluid) of the lungs. As little as 2.5 milligrams per litre in the atmosphere causes death within a few minutes, but less than 0.0001 percent by volume may be tolerated. Chlorine was the first gas used in chemical warfare in World War I. The gas is easily liquefied by cooling or by pressures of a few atmospheres at ordinary temperature.

Toxicity of chlorine

Chlorine has a high electronegativity and a high electron affinity, the latter being even slightly higher than that of fluorine. The affinity of chlorine for hydrogen is so great that the reaction proceeds with explosive violence in light, as in the following equation:

$$H_2 + Cl_2 \xrightarrow{h\nu} 2HCl.$$

In the presence of charcoal, the combination of chlorine and hydrogen takes place rapidly (but without explosion) in the dark. A jet of hydrogen will burn in chlorine with a silvery flame. Its high affinity for hydrogen allows chlorine to react with many compounds containing hydrogen. Chlorine reacts with hydrocarbons, for example, substituting chlorine atoms for the hydrogen atoms successively. If the hydrocarbon is saturated, however, chlorine atoms readily add to the double or triple bond.

Chlorine reacts with many elements of both metals and nonmetals to give chlorides. Only toward carbon, nitrogen, and oxygen is it fairly inert. The products of reaction with chlorine usually are chlorides with high oxidation numbers, such as iron(III) chloride ($FeCl_3$), tin(IV) chloride ($SnCl_4$), or antimony(V) chloride ($SbCl_5$), but it should be noted that the chloride of highest oxidation number of a particular element is frequently in a lower oxidation state than the fluoride of highest oxidation number. Thus, vanadium forms a pentafluoride, whereas the pentachloride is unknown, and sulfur gives a hexafluoride but no hexachloride. With sulfur, even the tetrachloride is unstable.

Chlorine displaces the less electronegative halogens from compounds. The displacement of bromides, for example, occurs according to the following equation:

$$2Br^- + Cl_2 \rightleftharpoons Br_2 + 2Cl^-.$$

Furthermore, it converts several oxides into chlorides. An example is the conversion of iron(III) oxide to the corresponding chloride:

$$2Fe_2O_3 + 6Cl_2 \rightleftharpoons 4FeCl_3 + 3O_2.$$

With carbon monoxide chlorine gives carbonyl chloride; and with sulfur dioxide, sulfuryl chloride:

$$CO + Cl_2 \rightleftharpoons COCl_2$$

$$SO_2 + Cl_2 \rightleftharpoons SO_2Cl_2.$$

Chlorine water

Chlorine is moderately soluble in water, yielding chlorine water, and from this solution a solid hydrate of ideal composition $Cl_2 \cdot 7.66H_2O$ is obtained. This hydrate is characterized by a structure that is more open than that of ice; the unit cell contains 46 molecules of water and six cavities suitable for the chlorine molecules. When the hydrate stands, disproportionation takes place—that is, one chlorine atom in the molecule is oxidized and the other is reduced. At the same time, the solution becomes acidic, as shown in the following equation:

$$\overset{0}{Cl_2} + H_2O \rightleftharpoons H\overset{+1}{Cl}O + H\overset{-1}{Cl}$$

in which the oxidation numbers are written above the atomic symbols. Chlorine water loses its efficiency as an oxidizing agent on standing, because hypochlorous acid gradually decomposes. The reaction of chlorine with alkaline solutions yields salts of oxo acids (see below).

The ionization potential of chlorine is high. Although ions in positive oxidation states are not very stable, high oxidation numbers are stabilized by coordination, mainly with oxygen and fluorine. In such compounds bonding is predominantly covalent, and chlorine is capable of exhibiting the oxidation numbers +1, +3, +4, +5, +6, and +7.

Analysis. Free chlorine may be recognized by its smell, its colour, and its characteristic reaction with mercury to produce white mercury(II) chloride. Tests for chloride ions are:

1. The formation of a white precipitate of silver chloride on addition of silver nitrate in dilute nitric acid. (This precipitate is soluble in the presence of ammonia.)
2. The formation of chromyl chloride, a red gas, by heating a solid sample with potassium dichromate and concentrated sulfuric acid. When chromyl chloride is passed into water, a yellow chromate solution forms (bromides and iodides do not form analogous compounds).
3. The evolution of free chlorine by heating the sample with manganese dioxide and concentrated sulfuric acid.

The following methods are available for the quantitative determination of free chlorine:

Quantitative determination of chlorine

1. The chlorine-containing gas is shaken with an aqueous solution of potassium iodide, and the resulting iodine is determined by titration.
2. Chlorine is reduced in alkaline solution by an alkali arsenite. Back-titration of excess arsenite is carried out with potassium bromate.
3. In the presence of an alkali hydroxide, chlorine is reduced to the chloride ion by hydrogen peroxide, and the excess alkali hydroxide is back-titrated with acid.
4. With sulfur dioxide or sodium thiosulfate, chlorine is reduced to chloride, and the latter is analyzed as silver chloride (see below).
5. Colorimetric measurements are carried out in the presence of *o*-toluidine in hydrochloric acid.

For the determination of chloride ions, one of the following methods may be recommended: (1) gravimetric analysis (analysis by weight of a given product) as silver chloride; (2) titration of neutral chloride solution with silver nitrate in the presence of potassium chromate; and (3) potentiometric titration (measurement of voltage changes) with silver nitrate, a process that can be carried out in the presence of bromide and iodide ions.

Most insoluble chlorides can be melted with soda, and the resulting melt is then usually soluble in water. Organic compounds containing chlorine are heated with alkali peroxide, and the product is dissolved in water.

Bromine. *Occurrence and distribution.* Because of the bad odour of the element, the French Academy of Science suggested the name bromine, from the Greek word *bromos,* meaning "bad smell," or "stench." Bromine occurs only in compounds and is considerably less abundant than chlorine. Apart from silver bromide (bromyrite), which is found in Mexico and Chile, the element is mainly found in seawater or in salt deposits. The bromide content of seawater is about 0.07 grams per litre, but the Dead Sea contains much more—up to five grams per litre. Natural salt deposits are the main source of bromine and its compounds in the United States. In Germany it is available as a by-product of the manufacture of potassium.

Bromine from seawater

Production and use. Bromine is produced on a large scale from seawater by treatment with chlorine in the presence of sulfuric acid, according to the following equation:

$$2Br^- + Cl_2 \rightleftharpoons Br_2 + 2Cl^-.$$

The product of the reaction is a dilute solution of bromine, from which the element is removed by blowing air through it. The free bromine is then mixed with sulfur dioxide, and the mixed gases are passed up a tower down which water is trickling. The following reaction takes place in the tower:

$$SO_2 + Br_2 + 2H_2O \rightarrow 2HBr + H_2SO_4,$$

resulting in a mixture of acids that is much richer in bromide ion than seawater. Treatment with chlorine again liberates bromine, which is freed from chlorine and purified by passage over moist iron filings. In Germany most bromine produced comes from the mother liquor obtained in the production of potassium salts. This mother liquor, which contains mainly magnesium bromide, is treated with chlorine in a continuous process to release the bromine.

Commercial bromine generally contains up to 0.3 percent chlorine. It is usually stored in glass bottles or, in the United States, in barrels coated with lead or Monel metal.

Much of the bromine produced is converted to ethylene dibromide ($C_2H_4Br_2$), which is added to gasoline with tetraethyl lead to prevent deposition of lead in the engine. Bromine is also used for the preparation of silver bromide, which is employed in photography, and in the production of catalysts, such as aluminum bromide, as well as of organic dyestuffs and various other organic compounds.

Physical and chemical properties. Free bromine is a reddish-brown liquid with an appreciable vapour pressure at room temperature. Both liquid bromine and the vapour are highly toxic and produce painful burns on the skin. Like the other halogens, bromine exists as diatomic molecules in all aggregation states.

Bromine water

About 3.41 grams of bromine dissolve in 100 millilitres of water at room temperature. The solution is known as bromine water. Like chlorine water, it is a good oxidizing agent, and it is more useful because it does not decompose so readily. It liberates free iodine from iodide-containing solutions and sulfur from hydrogen sulfide. Sulfurous acid is oxidized by bromine water to sulfuric acid. In the sunlight bromine water decomposes, with release of oxygen, as in the following equation:

$$H_2O + Br_2 \longrightarrow 2HBr + \tfrac{1}{2}O_2.$$

From bromine water a hydrate can be isolated that contains 172 water molecules and 20 cavities capable of accommodating the bromine molecules. Bromine dissolves in aqueous alkali hydroxide solutions, giving bromides, hypobromites, or bromates, depending on the temperature. Bromine is readily extracted from water by organic solvents such as carbon tetrachloride, chloroform, or carbon disulfide, in which it is very soluble. In the organic solvents it gives an orange solution.

The electron affinity of bromine is high and similar to that of chlorine. It is, however, a less powerful oxidizing agent, chiefly because of the weaker hydration of the bromide ion as compared with the chloride ion. Similarly, a metal–bromine bond is weaker than the corresponding metal–chlorine bond, and this difference is reflected in the chemical reactivity of bromine, which lies between that of chlorine and that of iodine.

Bromine combines violently with the alkali metals and with phosphorus, arsenic, and antimony but less violently with certain other metals. Bromine displaces hydrogen from saturated hydrocarbons and adds to unsaturated hydrocarbons, though not as readily as chlorine does.

The ionization potential of bromine is high, and compounds containing bromine in positive oxidation numbers are stabilized by appropriate ligands, mainly oxygen and fluorine. Compounds with the oxidation numbers +1, +3, +4, +5, and +7 are known; they all contain covalent bonds.

Analysis. A sensitive test for bromine is the reaction with fluorescein to give a deep red colour caused by bromination of the organic molecule, or by its reaction with fuchsine dyes in the presence of sulfurous acid, to give a deep blue colour. A more common test involves heating of the sample with dilute sulfuric acid in the presence of potassium dichromate; the bromine is then extracted with chloroform, and, upon addition of potassium iodide, the pink colour of iodine appears. The presence of bromine may also be recognized by the evolution of hydrogen bromide containing some brown bromine vapour when a solid sample is treated with concentrated sulfuric acid. Alternatively, chlorine may be added to an aqueous solution of a sample containing bromide, with development of a brown colour (free bromine).

For the quantitative determination of bromine the following methods are recommended:

$$Br_2 + 2(S_2O_3)^{2-} \rightleftharpoons (S_4O_6)^{2-} + 2Br^-.$$

1. Free bromine is titrated with sodium thiosulfate in the presence of potassium iodide:

$$Br^- + Ag^+ \rightleftharpoons AgBr.$$

2. Bromides may be determined either gravimetrically (by weight analysis) or by titration with silver nitrate:
3. In the presence of chloride and iodide, the potentiometric method may be used (as with chlorine).
4. In the absence of iodide, bromide may be oxidized to bromine, which is then determined in the distillate. Alternatively, bromide may be oxidized to bromate by hypochlorous acid. The excess of the oxidizing agent is destroyed by sodium formate, and iodine is liberated by addition of potassium iodide and acid, the free iodine being titrated by thiosulfate.
5. For the determination of bromine in an organic compound, the latter is oxidized by nitric acid, and the bromine is determined as silver bromide.

Iodine. *Occurrence and distribution.* Because of its violet-coloured vapours, the element was given the name iodine from the Greek word *ioeides,* "violet coloured." Iodine occurs to a small extent in seawater and is formed in seaweeds, oysters, and cod livers. Sodium iodate ($NaIO_3$) is contained in crude Chile saltpetre (sodium nitrate, $NaNO_3$). The human body contains iodine in the compound thyroxine, which is produced in the thyroid gland. It occurs in small quantities in much animal and vegetable matter.

Iodine from brines

Production and use. Iodine is produced commercially either from Chile saltpetre or from iodine-containing brines. In the former process, the salt is dissolved in hot water and the saltpetre allowed to crystallize on cooling. The mother liquor is used for further extractions until the extracts contain up to nine grams of iodine per litre. Sodium hydrogen sulfite is then added in order to reduce all iodate to iodide, and the solution is nearly neutralized with sodium carbonate. Fresh mother liquor is then added until all iodide is oxidized by the iodate to free iodine, according to the equation:

$$2NaIO_3 + 5NaHSO_3 \rightleftharpoons 3NaHSO_4 + 2Na_2SO_4 + H_2O + I_2.$$

The solid, containing up to 80 percent iodine, is collected, washed with water, and pressed into cheeselike blocks. These are heated to distill off both iodine and water.

Natural brines, or brines extracted from oil wells containing up to 150 milligrams of iodine per litre, are found in Java, California, and northern Italy. Impurities, such as clay, sand, and oil, are removed by filtration, and the solution is passed through a stream of sulfur dioxide and then through a number of containers containing bundles of copper wire. The copper(I) iodide that forms is removed by filtration, washed with water, dried, and finely ground. The product is heated with potassium carbonate to give potassium iodide, which is then oxidized to the free element with dichromate and sulfuric acid:

$$Cu_2I_2 + K_2CO_3 \rightleftharpoons 2KI + Cu_2O + CO_2$$

$$6KI + Na_2Cr_2O_7 + 7H_2SO_4 \rightleftharpoons 3I_2 + Cr_2(SO_4)_3 + Na_2SO_4 + 3K_2SO_4 + 7H_2O.$$

In an alternate process, chlorine is used as the oxidizing agent:

$$2KI + Cl_2 \rightleftharpoons 2KCl + I_2.$$

Iodine from seaweed

For a long time, iodine has been recovered on a commercial scale from seaweed. This is dried and burnt; the ash is leached with water; sodium sulfate and sodium chloride are removed by crystallization; and the remaining solution is concentrated by evaporation of water. The final solution, which contains 30 to 100 grams of iodine per litre, is treated with sulfuric acid in order to decompose any sulfite, and sulfide and manganese dioxide are added to release iodine, which is vapourized and purified by sublimation. Alternatively, addition of cupric sulfate gives cuprous iodide.

Iodine is widely used as a disinfectant and antiseptic, frequently in a solution of alcohol and water containing potassium iodide. Several compounds of iodine, such as iodoform (CHI_3), also serve as antiseptics.

Because iodine is converted to thyroxine in the thyroid gland, a small amount of iodine is essential for the body. In many places drinking water contains sufficient iodine for this purpose. In the absence of iodine in the water supply, however, goitre is prevalent, and a small quantity of iodine is frequently added to table salt in order to ensure against iodine deficiency.

Iodine and its compounds are used extensively in analytical chemistry, many analytical procedures being based on the release or uptake of iodine and its subsequent titration with sodium thiosulfate (iodometry). Unsaturation of fats (that is, the number of double bonds between carbon atoms) is determined by addition of free iodine (iodine number). Iodine compounds are also employed as catalysts in certain classes of organic reactions. Iodine, silver iodide, and potassium iodide are used in photography. Silver iodide is also used to seed clouds to induce rain. Iodine has been introduced into metallurgical processes for the production of certain transition metals in a high state of purity, among them titanium, zirconium, thorium, chromium, and cobalt. Electronic equipment, such as scintillation counters or neutron detectors, contain single-crystal prisms consisting of alkali metal iodides.

Solid iodine

Physical and chemical properties. Iodine is solid at room temperature, dark coloured, and with a glittering crystalline appearance. The molecular lattice contains discrete diatomic molecules, which are also present in the molten and the gaseous states. Above 700° C dissociation into iodine atoms becomes appreciable.

Iodine has a moderate vapour pressure at room temperature and in an open vessel slowly sublimes (sublimation is vaporization of a solid—comparable to distillation of a liquid). For this reason, iodine is best weighed in a stoppered bottle; for the preparation of an aqueous solution the bottle may contain a solution of potassium iodide, which considerably decreases the vapour pressure of iodine, a complex (triiodide) being readily formed.

Molten iodine may be used as a nonaqueous solvent for iodides. The electrical conductivity of molten iodine has in part been ascribed to the following self-ionization equilibrium:

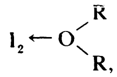
$$2I_2 \rightleftharpoons I^+ + I_3^-.$$

The alkali iodides are soluble in molten iodine and give conducting solutions typical of weak electrolytes. Alkali iodides react with compounds containing iodine with the oxidation number +1, such as iodine bromide, as in the following equation:

$$KI_3 + IBr \rightleftharpoons KBr + 2I_2.$$

In such reactions the alkali iodides may be regarded as bases.

The iodine molecule can act as a Lewis acid in that it combines with various Lewis bases. The interaction is weak, however, and few solid complex compounds have been isolated. The complexes are easily detected in solution and are referred to as charge-transfer complexes. Iodine, for example, is slightly soluble in water to give a yellowish-brown solution. Brown solutions are also formed with alcohol, ether, ketones, and other compounds acting as Lewis bases through an oxygen atom, as in the following example:



in which the R groups represent various organic groups.

Iodine gives a red solution in benzene, which is regarded as the result of a different type of charge-transfer complex. In inert solvents, such as carbon tetrachloride or carbon disulfide, violet-coloured solutions are obtained that contain uncoordinated iodine molecules. Iodine reacts also with iodide ions, because the latter can act as Lewis bases, and for this reason the solubility of iodine in water is greatly enhanced in the presence of an iodide. When cesium iodide is added, crystalline cesium triiodide may be isolated from the reddish-brown, aqueous solution. Iodine forms a blue complex with starch, and this colour test is used to detect small amounts of iodine.

The electron affinity of the iodine atom is not much different from those of the other halogen atoms. Iodine is a weaker oxidizing agent than bromine, however, because the iodide ion is a weaker Lewis base than the bromide ion and is hydrated to a smaller extent. In fact, the oxidizing properties of iodine are considerably weaker than those of bromine, chlorine, or fluorine. The following reaction—oxidation of arsenite, $(AsO_3)^{3-}$—in aqueous solution proceeds only in the presence of sodium hydrogen carbonate, which acts as a buffer:

$$(AsO_3)^{3-} + I_2 + 2OH^- \rightleftharpoons (AsO_4)^{3-} + 2I^- + H_2O.$$

In acidic solution, arsenate is reduced to arsenite, whereas, in strongly alkaline solution, iodine is unstable, and the reverse reaction occurs.

Volumetric determination of iodine

The most familiar oxidation by iodine is that of the thiosulfate ion, which is oxidized quantitatively to tetrathionate, as shown:

$$2(S_2O_3)^{2-} + I_2 \rightleftharpoons (S_4O_6)^{2-} + 2I^-.$$

This reaction is used to determine iodine volumetrically. The consumption of iodine at the end point is detected by the disappearance of the blue colour produced by iodine in the presence of a fresh starch solution.

Iodine combines directly with many elements. Silver and aluminum are easily converted into the respective iodides. White phosphorus unites readily with iodine, but no compounds of sulfur and iodine are known. The ionization potential of the iodine atom is considerably smaller than that of the lighter halogen atoms, and this is in accord with the existence of various compounds containing iodine in a positive oxidation number.

Analysis. Free iodine is detected (1) by the violet colour of the vapour or of its solution in carbon tetrachloride or carbon disulfide or (2) by the bright-blue colour produced in the presence of fresh starch solution in water, a very sensitive test.

Detection of iodide ions

Iodide ions may be detected in water (1) by the yellow precipitate of silver iodide, insoluble in water and ammonia solution, which is produced by addition of silver nitrate in the presence of dilute nitric acid, (2) by the formation of iodine on addition of chlorine or bromine water, (3) by the formation of iodine in the presence of other oxidizing agents, such as hydrogen peroxide or potassium dichromate, or (4) by the scarlet precipitate of mercury(II) iodide formed on addition of mercury(II) chloride, as in the equation below:

$$2I^- + HgCl_2 \rightleftharpoons HgI_2 + 2Cl^-.$$

This precipitate is dissolved by excess of iodide ions because of formation of a complex ion:

$$HgI_2 + 2I^- \rightleftharpoons (HgI_4)^{2-}.$$

Iodate or periodate is reduced by sulfurous acid to iodide and may be detected as such.

For the quantitative determination of iodine, one of the following methods may be recommended: (1) gravimetrically, by precipitation as silver iodide; (2) volumetrically, by titrating iodine with a standardized solution of sodium thiosulfate (using starch as an indicator); or (3) potentiometric titration with silver nitrate, which is applicable in the presence of both chloride and bromide. The second method is applied in the determination of many oxidizing substances. Dichromate, for example, reacts with excess potassium iodide in the presence of sulfuric acid, as shown:

$$(Cr_2O_7)^{2-} + 6I^- + 14H^+ \rightleftharpoons 2Cr^{3+} + 3I_2 + 7H_2O.$$

The iodine liberated is treated with standard thiosulfate solution.

Astatine. Because the element astatine has no stable or long-lived isotopes, it was given its name from the Greek word *astatos,* meaning "unstable." Minute amounts of the very short-lived isotopes, astatine-215, -218, and -219, occur in nature in radioactive equilibrium with certain long-lived, naturally occurring radioelements. In the body, astatine is concentrated in the thyroid gland. A substantial portion, however, distributes throughout the body, where it acts as an internal radiation source.

Preparation of astatine

Production and use. The only practical way of obtaining astatine is by synthesizing it through nuclear reactions. The first synthesis was effected in 1940 by the bombardment of bismuth with alpha particles to obtain astatine-211.

Astatine is usually prepared according to the general equation:

$$^{209}\mathrm{Bi}\,(a, xn)\,^{(213-x)}\mathrm{At},$$

which indicates that bismuth-209 takes up one alpha particle and emits x neutrons to form an isotope of astatine, whose atomic weight depends on the number of neutrons lost. Metallic bismuth may be used as a target material. From this, astatine may readily be removed by distillation in air from a stainless-steel tube. The free element begins to distill at 271° C (the melting point of bismuth), but the operation is best carried out at 800° C with subsequent redistillation. If an aqueous solution of astatine is desired, the element may be separated by washing with an appropriate aqueous solution. Alternatively, the halogen may be removed from the target by chemical methods, such as dissolving in nitric acid, the latter being removed by boiling.

Another procedure involves the use of a metallic thorium target, which—after bombardment—is dissolved in concentrated hydrochloric acid containing hydrogen fluoride and chlorine.

Use of iodine as a carrier in experiments

Analysis. With the exception of a few spectrographic and mass-spectrographic studies, all investigations of astatine chemistry have utilized tracer techniques at extreme dilution—concentrations around or below 10^{-10} molar, molarity being the number of moles (gram molecular weight) per litre of solution. At such concentration, the effects of impurities can be very serious, especially for a halogen such as astatine, which exists in several oxidation states and can form many organic compounds. Iodine has been used as a carrier in most experiments. Techniques applied include coprecipitation, solvent extraction, ion exchange, and other forms of chromatography (separation by adsorption differences), electrodeposition (deposition by an electric current), electromigration (movement in an electric field), and diffusion. A direct identification of some astatine compounds has been made by mass spectrography.

Except for nuclear properties, the only physical property of astatine to be measured directly is the spectrum of atomic astatine. Other physical properties have been predicted from theory and by extrapolation from the properties of other elements.

Chemical properties. The astatide ion, At^-, is quantitatively coprecipitated with insoluble iodides, such as silver iodide or thallium(I) iodide. The diffusion coefficient of the iodide ion is 1.42 times that of the astatide ion, which moves more slowly toward the anode than the former under given conditions. The ion is formed by reduction of the element, using, for example, zinc and acid. It is oxidized to the zero valence state by ferric ion, Fe^{3+}, iodine (I_2), and dilute nitric acid. Thus, the astatide ion is a stronger reducing agent than the iodide ion, and free iodine is a stronger oxidizing agent than astatine.

Free astatine is characterized by volatility from solution and by extractability into organic solvents. It undergoes disproportionation in alkaline media. Astatine is coprecipitated with cesium iodide and thus appears to form polyhalide anions. Astatine extracted into chloroform has been shown to coprecipitate homogeneously with iodine when a portion of the latter is crystallized. Astatine seems to be present as the iodide, which appears to be more polar (*i.e.,* showing separation of electrical charge) in character than iodine bromide.

Astatine is known to occur in positive oxidation numbers. The astatate ion, $(AtO_3)^-$, is coprecipitated with insoluble iodates, such as silver iodate $(AgIO_3)$, and it is obtained by oxidation of lower oxidation states with hypochlorite, periodate, or persulfate. So far no evidence for perastatate has been found, but this may be because the ion, $(AtO_6)^{5-}$, may show little tendency to coprecipitate with potassium iodate (KIO_4).

Astatine in the $+1$ state is stabilized by complexation, and complexes formulated as dipyridine astatine(I) perchlorate $[At(py)_2]$ $[ClO_4]$ and dipyridine astatine(I) nitrate $[At(py)_2]$ $[NO_3]$ have been prepared. Compounds with the formulas $(C_6H_5)AtCl_2$, $(C_6H_5)_2AtCl$, and $(C_6H_5)AtO_2$ have also been obtained. A variety of methods may be used to synthesize astatobenzene, C_6H_5At. (V.G./Ed.)

Noble gases

The so-called noble gases consist of a group of six chemical elements that exist under ordinary conditions as colourless, odourless, tasteless, nonflammable gases: helium (He), neon (Ne), argon (Ar), krypton (Kr), xenon (Xe), and radon (Rn). They traditionally have been labelled Group 0 in the periodic table (see Figure 10) because for decades after their discovery it was believed that they had a valence of zero; that is, that their atoms could not combine with those of other elements to form chemical compounds. Their electronic structures and the finding that some of them do indeed form compounds suggest that a more appropriate designation would be Group VIIIa.

When the members of the group were discovered and identified they were thought to be exceedingly rare, as well as chemically inactive, and therefore were called the rare gases or the inert gases. It is now known, however, that several of these elements are quite abundant on Earth and in the rest of the universe, so the designation rare is misleading. Similarly, use of the term inert has the drawback that it often is applied to gases such as nitrogen and carbon dioxide to connote their nonflammability. In chemistry and alchemy, the word noble long has signified the passivity toward oxygen of a group of metals, such as gold and platinum; it applies in the same sense to the group of gases covered here.

The abundances of the noble gases decrease as their atomic numbers increase; helium is, in fact, the most plentiful element in the universe except hydrogen. All the noble gases are present in the Earth's atmosphere and, except for helium and radon, their major commercial source is the air, from which they are obtained by liquefaction and fractional distillation. Most helium is produced commercially from certain natural gas wells. Radon usually is isolated as a product of the radioactive decomposition of dissolved radium compounds (the nuclei of radium atoms spontaneously decay by emitting energy and particles; the particles are the nuclei of helium and radon atoms).

Important industrial applications

Several important uses of the noble gases rest on their marked lack of chemical reactivity. Their indifference toward oxygen, for example, confers utter nonflammability upon all six noble gases. Although helium is not quite as buoyant as hydrogen, its incombustibility makes it a safer lifting gas for lighter-than-air aircraft. The noble gases—most often helium and argon, the least expensive—are used to provide chemically unreactive environments for such operations as cutting, welding, and refining of metals (atmospheric oxygen and, in some cases, nitrogen or carbon dioxide would react with the hot metal) and in the handling of other easily attacked materials.

The noble gases absorb and emit electromagnetic radiation in a much less complex way than do other substances. The absorption and emission behaviour is utilized in the employment of these gases in discharge lamps and fluorescent lighting devices: if any of them is confined at low pressure in a glass tube and an electrical discharge is passed through it, the gas glows. Neon produces the familiar orange-red colour of advertising signs; xenon emits a beautiful blue.

The very low boiling points and melting points of the noble gases make them useful as refrigerants in the study of matter at extremely low temperatures. The low solubility of helium in fluids leads to its use in admixture with oxygen for breathing by deep-sea divers: because helium does not dissolve in the blood, it does not form bubbles upon decompression (as nitrogen does, leading to the condition known as bends). Xenon has been used as an anesthetic; although it is costly, it is nonflammable and readily eliminated from the body. Radon is highly radioactive; its only uses have been those that exploit this property, as, for example, in radiotherapy.

The compounds of the noble gases are powerful oxidizing agents (substances that tend to remove electrons from others) and have potential value as reagents in the synthesis of chemical compounds.

HISTORY

Discovery of argon

In 1785 Henry Cavendish, an English chemist and physicist, found that air contains a small proportion (slightly less than 1 percent) of a substance that is chemically less active than nitrogen. A century later Lord Rayleigh, an English physicist, isolated from the air a gas that he thought was pure nitrogen but found that it was denser than nitrogen prepared chemically by liberating it from its compounds. He reasoned that his aerial nitrogen must contain a small amount of a denser gas. Sir William Ramsay, a British chemist, collaborated with Rayleigh in isolating (1894) this gas, which proved to be a new element, argon.

After the discovery of argon, and at the instigation of other scientists, Ramsay undertook to investigate the gas evolved upon heating the mineral clevite, which was thought to be a source of argon. The gas proved (1895) instead to be helium, which in 1868 had been detected spectroscopically in the Sun but had not been found on Earth. Ramsay and his co-workers searched for related gases and by fractional distillation of liquid air discovered krypton, neon, and xenon, all in 1898. Radon, first identified in 1900, was established as a member of the noble-gas group in 1904.

In 1895 the French chemist Henri Moissan failed in an attempt to bring about a reaction between fluorine and argon. In fact, all late 19th- and early 20th-century efforts to prepare chemical compounds of argon failed. The lack of chemical reactivity implied by these failures was of significance in the development of theories of atomic structure. In 1913 the Danish physicist Niels Bohr proposed that the electrons in atoms are arranged in successive shells having characteristic energies and capacities and that the capacities of the shells determine the numbers of elements in the periods of the periodic table. On the basis of experimental evidence relating chemical properties to electron distributions, it was suggested that in the atoms of each noble gas beyond helium the electrons are arranged in these shells in such a way that the outermost shell always contains eight electrons, no matter how many others (in the case of radon, 78 others) were arranged inside those of eight.

Octet theory

In a theory of chemical bonding advanced by others in 1916, this octet of electrons was taken to be the most stable arrangement for the outermost shell of any atom. Although only the noble-gas atoms possessed this arrangement, it was the condition toward which the atoms of all other elements tended in their chemical bonding. Certain elements satisfied this tendency by either gaining or losing electrons outright, becoming ions; other elements shared electrons, forming stable combinations linked together by covalent bonds. The valences of elements—that is, the proportions in which their atoms combined to form ionic or covalent compounds—were thus controlled by the behaviour of their outermost electrons, which for this reason were called the valence electrons. This theory rationalized the chemical bonding of the reactive elements as well as the relative inactivity of the noble gases, which is their chief characteristic. (Additional information appears in the article MOLECULES.)

The outer electrons of the atoms of the heavier noble gases, screened from the nucleus by intervening electrons, are held less firmly and can be removed more easily from the atoms than can the electrons of the lighter noble gases. This fact had been known for a long time from experiments using electrical and magnetic fields; the energy required for removal of one electron is called the first ionization potential. In 1962 it was discovered that platinum hexafluoride would oxidize molecular oxygen to form a salt. Knowledge that the first ionization potential of xenon is very close to that of oxygen led to the suggestion that a salt of xenon might be formed similarly. In the same year it was established that it is indeed possible to remove electrons from xenon by chemical means—that is, to oxidize xenon—when two teams of chemists independently prepared fluorides of that element. This achievement was followed by the preparation of other xenon compounds and of the fluorides of radon (1962) and krypton (1963).

GENERAL PROPERTIES OF THE GROUP

Each noble-gas element is situated in the periodic table between an element of the most electronegative group, the halogens (Group VIIa, the atoms of which add electrons to achieve the octet and thereby become negative ions), and an element of the most electropositive group, the alkali metals (Group Ia, the atoms of which lose electrons to become positive ions). The noble gases thus form a dramatic transition group in the periodic table of the elements and are neither electropositive nor electronegative but, relatively speaking, neutral, neither gaining nor losing electrons easily.

The sizes of the atoms of noble gases increase smoothly with the increase in atomic number, partially because repulsion between the electrons prevents them from occupying the same region of space and forces each successive shell to extend into a larger concentric spherical volume. In the largest atoms, the weaker attraction of the nucleus for the electrons in the most distant shell explains the greater ease with which the outer electron clouds of these atoms are distorted by other charged bodies, or polarized. A polarized atom as a whole remains electrically neutral but the distribution of charge within it becomes unsymmetrical, the centre of gravity, so to speak, of the negative charge being displaced from that of the positive charge. The polarizability of the noble gases increases markedly through the group from helium to radon as the ionization potential decreases.

Clathrates

Generally speaking, a polarizable substance is attracted more strongly to surfaces and dissolves in fluids more extensively than a nonpolarizable one. Because the polarizability of helium is minimal, it shows little or no tendency to adsorb upon surfaces and it is not very soluble in fluids. These properties may account for the fact that helium does not induce narcosis when it is breathed by divers at high pressures and that it does not form clathrates. (Clathrates—from a Latin word meaning "enclosed by a lattice"—are substances in which the molecules of one compound, called the host, form a cagelike crystalline lattice within which there are open spaces that may be occupied by molecules of a second compound or element, called the guest, although no chemical bonds are formed between the host and the guest. The stability of a clathrate is influenced by the closeness with which the guest particles fit into the holes as well as by the polarizability of the guest.) The properties of small size and low polarizability make helium a highly mobile gas; *i.e.,* one that is not trapped by various structures of other atoms. The greater polarizability of argon and the heavier noble gases accounts for their formation of clathrates, including those in which water is the host—hydrates—and also accounts for the ready adsorption of these gases on surfaces. The narcotic activity of the noble gases also parallels their polarizability. The noble gases reduce the sensitivity of living cells to attack by oxygen under the influence of radiation; the heavier gases are more effective than the lighter ones in this action, perhaps simply as a result of their greater solubility in fluids, but possibly also as a consequence of their greater ability to absorb neutrons.

INDIVIDUAL NOBLE GASES

Helium. Because the first evidence of the existence of helium was the presence of certain wavelengths in the yellow region of the spectrum of the light emitted from the Sun, the name of the element was derived from the Greek word *hēlios,* "sun." Helium constitutes about 23 percent of the mass of the universe, but only about eight parts per 1,000,000,000 of the Earth's crust; ordinary air contains about five parts per 1,000,000 of helium.

The nucleus of every helium atom contains two protons but, as is the case with all elements, isotopes of helium exist. The known isotopes of helium contain from one to six neutrons, so their mass numbers range from three to eight. Of these six isotopes, only the two having mass numbers of three (helium-3, symbolized ^{3}He) and four (helium-4, ^{4}He) are stable; all the others are radioactive, decaying very rapidly into other substances. Helium-4 is by far the more plentiful of the stable isotopes: helium-4 atoms outnumber those of helium-3 about 700,000 to one in atmospheric helium and about 7,000,000 to one in certain helium-bearing minerals.

Most of the helium of commerce is derived from natural gas produced in the southwestern United States. The helium, which comprises between 1.5 and 7 percent of the gas, is separated by continuous processes involving several steps. A portion of the helium market is supplied from plants that liquefy air on a large scale; the amount of helium obtainable from 1,000 tons of air is about 112 cubic feet, measured at room temperature and atmospheric pressure.

Liquid and solid helium

The boiling and freezing points of helium are lower than those of any other known substance. Helium is the only element that cannot be solidified by sufficient cooling at normal atmospheric pressure; it is necessary to apply pressure of 25 atmospheres at a temperature of 1 K (−272.15° C) to convert it to its solid form.

The isotope helium-4 is unique in having two liquid forms. The form designated I exists at temperatures from its boiling point of 4.21 K down to 2.18 K (−270.97° C); form II exists at still lower temperatures. Liquid helium II exhibits the property called superfluidity: its viscosity, or resistance to flow, is so low that it has not been measured. This liquid spreads in a thin film over the surface of any substance it touches, and this film flows without friction even against the force of gravity.

A liquid mixture of the two isotopes helium-3 and helium-4 separates at temperatures below about 0.8 K into two layers. One layer is practically pure helium-3; the other is mostly helium-4, but retains about 6 percent helium-3 even at the lowest temperatures achieved. Dissolving helium-3 in helium-4 is accompanied by a cooling effect that has been utilized in the construction of cryostats (devices for production of very low temperatures) that can attain—and maintain for periods of days—temperatures as low as 0.01 K.

Neon. Neon, the name of which is derived from the Greek word *neos,* "new," is present in some minerals, but its only commercial source is the atmosphere, of which it comprises 18 parts per 1,000,000 by volume. Because its boiling point is −246° C, neon remains, along with helium and hydrogen, in the small fraction of air that resists liquefaction upon cooling to −196° C. Neon is isolated from this cold, gaseous mixture by bringing it into contact with activated charcoal, which adsorbs the neon and hydrogen; removal of hydrogen is effected by adding enough oxygen to convert it all to water, which, along with any surplus oxygen, condenses upon cooling.

Neon was the first element shown to consist of more than one stable isotope. In 1913 application of the technique of mass spectrometry revealed the existence of neon-20 and neon-22, which comprise 90.92 and 8.82 percent, respectively, of the naturally occurring mixture. The third stable isotope, neon-21, which makes up 0.26 percent of natural neon, was detected later. Five radioactive isotopes of neon also have been identified.

Argon. The first of the noble gases to be discovered, argon was named from the Greek word *argos,* "lazy," because of its chemical inertness. In cosmic abundance, argon ranks approximately 12th among the 100-odd chemical elements; although the stable isotopes argon-36 and argon-38 make up all but a trace of this element in the universe, the third stable isotope, argon-40, comprises 99.60 percent of the argon found on Earth. The terrestrial preponderance of argon-40 presumably arose from the formation of this isotope by the radioactive decay of potassium-40.

Table 22: Some Properties of the Noble Gases

	helium	neon	argon	krypton	xenon	radon
Atomic number	2	10	18	36	54	86
Atomic weight	4.003	20.182	39.948	83.797	131.295	(222)*
Melting point						
(°C)	−272.15†	−248.67	−189.2	−156.6	−111.9	−71
(°K)	1	24.48	83.95	116.55	161.25	202
Boiling point						
(°C)	−268.94	−246.05	−185.88	−152.3	−107.10	−61.8
(°K)	4.21	27.10	87.27	120.8	166.05	211.07
Density at 0° C, 1 atm (g/1)	0.17847	0.89994	1.78403	3.733	5.8811	9.73
Solubility in water at 20° C (cu cm of gas per 1,000 g water)	8.61	10.5	33.6	59.4	108.1	230
Electronic configuration	$1s^2$	$1s^22s^22p^6$	(Ne)$3s^23p^6$	(Ar)$3d^{10}4s^24p^6$	(Kr)$4d^{10}5s^25p^6$	(Xe)$4f^{14}5d^{10}6s^26p^6$
Isotopic abundance (terrestrial, percent)	^{3}He (0.00013) ^{4}He (100)	^{20}Ne (90.92) ^{21}Ne (0.257) ^{22}Ne (8.82)	^{36}Ar (0.337) ^{38}Ar (0.063) ^{40}Ar (99.600)	^{78}Kr (0.354) ^{80}Kr (2.27) ^{82}Kr (11.56) ^{83}Kr (11.55) ^{84}Kr (56.90) ^{86}Kr (17.37)	^{124}Xe (0.096) ^{126}Xe (0.090) ^{128}Xe (1.912) ^{129}Xe (26.44) ^{130}Xe (4.08) ^{131}Xe (21.18) ^{132}Xe (26.89) ^{134}Xe (10.44) ^{136}Xe (8.87)	^{219}Rn (trace) ^{220}Rn (trace) ^{222}Rn (trace)
Radioactive isotopes (mass numbers)	6	17–19, 23, 24	33, 35, 37, 39, 41, 42	74–77, 79, 81, 85, 87–94	118–123, 125, 127, 133, 135, 137–142	204–213 215–224
Colour of light emitted by gaseous discharge tube	yellow	red	red or blue	yellow green	blue to green	—
Heat of fusion (cal/mole)	5 (3.5° K, 100 atm)	80.1	280.8	390.7	548.5	—
Heat of vaporization (cal/mole)	19.4	414	1,557.5	2,158	3,020	4,325
Specific heat (cal/mole/°C)	4.9680	4.9680	4.9680	4.9680	4.9680	4.9680
Critical temperature (°K)	5.25	44.5	150.85	209.35	289.74	378.15
Critical pressure (atm)	2.26	26.9	48.3	54.3	57.64	62
Critical density (g/cm³)	0.0693	0.484	0.536	0.908	1.100	—
Thermal conductivity at 0° C, 1 atm (cal/cm-sec-°C)	33.90×10^{-5}	11.00×10^{-5}	3.920×10^{-5}	2.09×10^{-5}	1.21×10^{-5}	—
Magnetic susceptibility (cgs units/mole)	−0.0000019	−0.0000072	−0.0000194	−0.000028	−0.000043	—
Crystal structure‡	hcp	fcc	fcc	fcc	fcc	fcc
Radius						
Atomic (Å)	1.3	1.6	1.92	1.98	2.18	—
Covalent (crystal) estimated (Å)	0.4–0.6	0.7	0.95	1.10	1.30	1.5–2.1
Static polarizability (Å³)	0.204	0.392	1.63	2.465	4.01	—
Ionization potential (first, eV)	24.586	21.563	15.759	13.999	12.129	10.747
Electronegativity (Pauling)	4.5	4.0	2.9	2.6	2.25	2.00

*Stablest isotope. †At 25.05 atm. ‡hcp = hexagonal close-packed, fcc = face centred cubic (cubic close-packed).

Argon is the most plentiful of the noble gases on Earth, comprising 0.934 percent by volume, or 1.288 percent by weight, of the atmosphere. The element is obtained from air by liquefaction and fractional distillation; although the boiling points of argon, oxygen, and nitrogen all lie within a few degrees of each other, efficient processing provides each gas in a purity of more than 99.9 percent.

Krypton. Krypton is named from the Greek word *kryptos,* "hidden." Traces of krypton are present in minerals and meteorites, but the usual commercial source is the atmosphere, which contains 1.14 parts per 10^6 by volume. Krypton also is formed by the nuclear fission of uranium triggered by slow neutrons: this source may be expected to become increasingly important because of the growing number of fission-power plants. Krypton has isotopes of every mass number from 74 through 95; six, with mass numbers 78, 80, 82, 83, 84, and 86, are stable. After it has been stored a few days, krypton obtained by nuclear fission contains only one radioactive isotope, krypton-85, which has a half-life of about 10 years, because all the other radioactive isotopes have half-lives of three hours or less. (The half-life is the length of time during which one-half of any original amount of an unstable substance decays.)

Because its boiling point is about 30° C higher than those of the major constituents of air, krypton is readily separated from liquid air by fractional distillation; it accumulates along with xenon in the least volatile portion. These two gases are further purified by adsorption onto silica gel, redistillation, and passage over hot titanium metal, which removes all impurities except other noble gases.

Krypton is the lightest of the noble gases that has been converted into chemical compounds.

Principal source

Xenon. The name xenon is derived from the Greek word *xenos,* "strange" or "foreign." Like several other noble gases, xenon is present in meteorites and certain minerals, but its only useful source has been the atmosphere, of which it composes 86 parts per 10^9 by volume. Nuclear reactors may become an important source of xenon, because the fission of uranium produces several isotopes of xenon.

The mass numbers of the known isotopes of xenon range from 118 to 144; nine of these numbers correspond to stable isotopes. The xenon isotopes produced in the greatest amount by nuclear fission are xenon-131, -132, -134, and -136, which are stable, and xenon-133, which is radioactive, having a half-life of 9.2 hours.

Xenon is the least volatile of the noble gases obtainable from the ir. Its purification has been mentioned above (see *Krypton*). Numerous compounds of xenon have been prepared since the discovery of the noble-gas compounds in 1962. Although xenon itself is an unusually safe anesthetic, its compounds appear to be toxic.

Radon. Radon was originally called radium emanation: it is the radioactive gas formed, along with helium, as radium decays. The term radon sometimes has been restricted to the isotope of mass 222, other isotopes being referred to as thoron (symbolized Tn, now called radon-220), formed from thorium; and actinon (An, now radon-219), formed from actinium. The name emanation, symbolized Em, has been used to denote the entire family now recognized as radon isotopes. Radon has no stable isotope; radioactive isotopes having masses ranging from 204 through 224 have been identified, the longest-lived of these being radon-222, which has a half-life of 3.82 days. All the isotopes decay, the stable end-products being helium and isotopes of heavy metals, in most cases lead. (Ed.)

Transition elements

The so-called transition elements constitute a group of 56 (the majority of the 107 known elements); while the term transition has no particular chemical significance, it is a convenient name by which to distinguish the similarity of the atomic structures and resulting properties of the elements so designated. They occupy the middle portions of the long periods of the periodic table of elements (see Figure 10) between the groups on the left-hand side and the groups on the right. Specifically, they form Groups IIIb, IVb, Vb, VIb, VIIb, VIII, and Ib. Of these 56 transition elements only 24 are treated here (see Table 23 for list of names), the others being dealt with in subsequent sections as explained below.

GENERAL PROPERTIES OF THE GROUP

The most striking similarities shared by the 24 elements in question are that they are all metals and that most of them are hard, strong, and lustrous, have high melting and boiling points, and are good conductors of heat and electricity. The range in these properties is considerable, therefore the statements are comparative with the general properties of all the other elements.

Table 23: Abundance of the Transition Elements in the Earth's Solid Crust
(in grams per ton)

Titanium	5.7×10^2	Ruthenium	0.001
Vanadium	135	Rhodium	0.001
Chromium	102	Palladium	0.01
Manganese	950	Silver	0.07
Iron	5.63×10^3	Hafnium	3.0
Cobalt	25	Tantalum	2.0
Nickel	75	Tungsten	1.5
Copper	55	Rhenium	0.001
Zirconium	165	Osmium	0.001
Niobium	20	Iridium	0.001
Molybdenum	1.5	Platinum	0.005
Technetium	~0	Gold	0.004

Many of the elements are technologically important: titanium, iron, nickel, and copper, for example, are used structurally and in electrical technology. Second, the transition elements form many useful alloys, with one another and with other metallic elements. Third, most of these elements dissolve in mineral acids although a few, such as platinum, silver, and gold, are called "noble"—that is, are unaffected by simple (nonoxidizing) acids.

Without exception, the elements of the main transition series (*i.e.,* excluding the lanthanides and actinides as specified below) exhibit variable valence and form stable compounds in two or more formal oxidation states (valence and oxidation states are explained in the following section).

The transition elements may be subdivided into two main types according to the electronic structures of their atoms, which are simply indicated here but fully explained in the section immediately following.

The three main transition series are called the first, second, and third transition series. The first series begins with either scandium (symbol Sc, atomic number 21) or titanium (symbol Ti, atomic number 22) and ends with copper (symbol Cu, atomic number 29). The second series includes the elements yttrium (symbol Y, atomic number 39) to silver (symbol Ag, atomic number 47). The third series extends from lanthanum (symbol La, atomic number 57) to gold (symbol Au, atomic number 79). These three main transition series are included in the set of 27 elements often called the *d*-block transition elements. Because scandium, yttrium, and lanthanum actually do not form compounds analogous to those of the other transition elements and because their chemistry is quite homologous to that of the lanthanides, they are excluded from the present discussion of the main transition elements. The *d*-block transition elements and some of their characteristic properties are listed in Table 24.

The lanthanides and the actinides comprise two other categories of transition elements. The first of these inner transition series includes the elements from cerium (symbol Ce, atomic number 58) to lutetium (symbol Lu, atomic number 71). These elements are called the lanthanides because the chemistry of each closely resembles that of lanthanum. Lanthanum itself is often regarded as one of the lanthanides. The actinide series consists of 15 elements from actinium (symbol Ac, atomic number 89) to lawrencium (symbol Lr, atomic number 103). These inner transition series are covered in the sections *Rare-earth elements* and *Actinide elements.* For elements 104 through 106, see below *Transuranium elements.*

(F.A.C./Ed.)

Electronic structure and position in the periodic table. The relative locations of the transition elements in the periodic table and their chemical and physical properties can best be understood by considering their electronic structures and the way in which those structures vary as atomic numbers increase.

Atomic orbitals. As noted earlier, the electrons associated with an atomic nucleus are localized, or concentrated, in various specific regions of space called atomic orbitals, each of which is characterized by a set of symbols (quantum numbers) that specify the volume, the shape, and orientation in space relative to other orbitals. An orbital may accommodate no more than two electrons. The en-

Table 24: Some Properties of the Transition Elements

	symbol	atomic number	atomic mass	density g cm^{-3}, 20° C	melting point °C	boiling point °C	type of crystal packing*	electrical resistivity microohm-cm	heat of atomization, at 298°, kJmole^{-1}	1st ionization potential electron volts
1st main series										
Titanium	Ti	22	47.90	4.507	1,668	3,260	hcp, bcc	42(0°)	473	6.82
Vanadium	V	23	50.94	6.11	1,890	3,000	bcc	24.8(20°)	515	6.74
Chromium	Cr	24	52.01	7.19	1,875	2,199	bcc, hcp	12.9(20°)	397	6.763
Manganese	Mn	25	54.938	7.44	1,244(±3)	2,097	complex	185(20°)	281	7.432
Iron	Fe	26	55.847	7.873	1,536(±1)	3,000	ccp, bcc	9.71(20°)	416	7.90
Cobalt	Co	27	58.94	8.90	1,493	3,100	ccp, hcp	5.68(0°)	425	7.86
Nickel	Ni	28	58.71	8.908	1,453	2,730	ccp, hcp	6.84(20°)	430	7.633
Copper	Cu	29	63.54	8.94	1,083	2,582	ccp	1.68(20°)	339	7.724
2nd main series										
Zirconium	Zr	40	91.22	6.506	1,850	4,377	bcc, hcp	40.0(0°)	611	6.984
Niobium	Nb	41	92.91	8.58	2,468(±10)	4,927	bcc	15.22(0°)	774	6.88
Molybdenum	Mo	42	95.94	10.22	2,610	5,560	bcc, hcp	7.2(0°)	659	7.10
Technetium	Tc	43	†	11.49	2,170	5,030	hcp	—	649	7.28
Ruthenium	Ru	44	101.1	12.45	2,310(±20)	4,080(±100)	hcp, ccp	6.71(0°)	669	7.364
Rhodium	Rh	45	102.91	12.41	1,960	3,700	ccp	4.33(0°)	577	7.46
Palladium	Pd	46	106.4	12.02	1,552	2,020	ccp	9.93(0°)	381	8.33
Silver	Ag	47	107.870	10.5	960.8	2,210	ccp	0.616(0°)	286	7.574
3rd main series										
Hafnium	Hf	72	178.50	13.29	2,230	5,200	hcp, bcc	35.5(20°)	703	7.9‡
Tantalum	Ta	73	180.95	16.65	2,996	5,427	bcc	13.6(0°)	781	7.88
Tungsten	W	74	183.85	19.3	3,410	5,930	bcc, complex	5.5(20°)	837	7.98
Rhenium	Re	75	186.22	21.04	3,170	5,630	hcp	19.14(0°)	791	7.87
Osmium	Os	76	190.2	22.61	3,050(±30)	5,020(±100)	hcp, ccp	8.12(0°)	728	8.7
Iridium	Ir	77	192.2	22.65	2,443	4,500	ccp	4.71(0°)	690	9
Platinum	Pt	78	195.09	21.45	1,769.3	4,100	ccp	9.85(0°)	566	9.0
Gold	Au	79	196.967	19.32	1,063	2,808	ccp	2.06(0°)	368	9.22

*hcp—hexagonal close-packed; ccp—cubic close-packed; bcc—body-centred cubic. †Does not occur in nature. The only isotope that has been obtained on a macroscopic scale has a mass number of 99. ‡Estimated value; nearly all are estimated or extrapolated values.

ergy involved in the interaction of an electron with the nucleus is determined by the orbital that it occupies, and the electrons in an atom distribute themselves among the orbitals in such a way that the total energy is minimum. Thus, by electronic structure, or configuration, of an atom is meant the way in which the electrons surrounding the nucleus occupy the various atomic orbitals available to them. The simplest configuration is the set of one-electron orbitals of the hydrogen atom. The orbitals can be classified, first, by principal quantum number, and the orbitals have increasing energy as the principal quantum number increases from 1 to 2, 3, 4, etc. (The sets of orbitals defined by the principal quantum numbers 1, 2, 3, 4, etc., are often referred to as shells designated *K, L, M, N,* etc.) For principal quantum number 1 there is but a single type of orbital, called an *s* orbital. As the principal quantum number increases, there are an increasing number of different types of orbitals, or subshells, corresponding to each: *s, p, d, f, g,* etc. Moreover, the additional orbital types each come in larger sets. Thus, there is but one *s* orbital for each principal quantum number, but there are three orbitals in the set designated *p,* five in each set designated *d,* and so on. For the hydrogen atom, the energy is fully determined by which orbital the single electron occupies. It is especially notable that the energy of the hydrogen atom is determined solely by the principal quantum number of the orbital occupied by the electron (except for some small effects that are not of concern here); that is, in hydrogen, the electron configurations of the third shell, for example, are equi-energic (of the same energy, whichever one the electron occupies), which is not the case with any of the other atoms, all of which contain two or more electrons.

To understand the electron configurations of other atoms, it is customary to employ the *Aufbau* (German: "building up") principle, the basis of which is that, to achieve a multi-electron configuration, the required number of electrons must be added to the orbitals one at a time, filling the most stable orbitals first, until the total number has been added. There is one restriction upon this conceptualization, namely, the Pauli exclusion principle, which states that only two electrons may occupy each orbital. Thus there can be no more than two electrons in any *s* orbital, six electrons in any set of *p* orbitals, ten electrons in any set of *d* orbitals, etc. In carrying out this process, however, one cannot simply use the ordering of electron orbitals that is appropriate to the hydrogen atom. As electrons are added they interact with each other as well as with the nucleus, and as a result the presence of electrons in some orbital causes the energy of an electron entering another orbital to be different from what it would be if this electron were present alone. The overall result of these interelectronic interactions (sometimes referred to as shielding) is that the relative order of the various atomic orbitals is different in many-electron atoms from that in the hydrogen atom; in fact, it changes continuously as the number of electrons increases.

Shielding

As multi-electronic atoms are built up, the various subshells *s, p, d, f, g,* etc. of a principal quantum number cease to be equi-energic; they all drop, although not by equal amounts, to lower energies. Overall lowering of energy occurs because the shielding from the nuclear charge that an electron in a particular orbital is given by all of the other electrons in the atom is not sufficient to prevent a steady increase in the effect that the charge in the nucleus has on that electron as the atomic number increases. In other words, each electron is imperfectly shielded from the nuclear charge by the other electrons. In addition the different types of orbitals in each principal shell, because of their different spatial distributions, are shielded to different degrees by the core of electrons beneath them; accordingly, although all of them decrease in energy, they decrease by different amounts, and thus their relative order in energy continuously changes. In order to specify the electron configuration of a particular atom, it is necessary to use the order of orbitals appropriate to the specific value of the atomic number of that atom. The behaviour of the various *d* and *f* orbitals is to be especially noted in regard to where the transition elements occur in the periodic table.

The argon atom (atomic number 18) has an electron configuration $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^6$ (*i.e.,* it has two electrons in the *s* orbital of the first shell; two in the *s* and six in the *p* orbitals of the second shell; two in the *s* and six in the *p* orbitals of the third shell: this expression often is abbreviated [Ar] especially in specifying the configurations of elements between argon and krypton, because it represents a common part of the configurations of all these elements). The 3*d* orbitals are more shielded from the nuclear charge than is the 4*s* orbital, and, consequently, the latter orbital has lower energy. The next electrons to be added enter the 4*s* orbital in preference to the 3*d* or 4*p* orbitals. The two elements following argon in the periodic table are potassium, with a single 4*s* electron, and calcium, with two 4*s* electrons. Because of the presence of the 4*s* electrons, the 3*d* orbitals are less shielded than the 4*p* orbitals; therefore, the first regular transition series begins at this point with the element scandium, which has the electron configuration [Ar] $4s^2 3d^1$. Through the next nine elements, in increasing order of atomic number, electrons are added to the 3*d* orbitals until, at the element zinc, they are entirely filled and the electron configuration is [Ar] $3d^{10} 4s^2$. The 4*p* orbitals are then the ones of lowest energy, and they become filled through the next six elements, the sixth of which is the next noble gas, krypton, with the electron configuration $1s^2\ 2s^2\ 2p^6\ 3s^2\ 3p^6\ 4s^2\ 3d^{10}\ 4p^6$, or [Kr].

First transition series

Throughout the next period the pattern of variation of the orbital energies is similar to that immediately preceding. When the configuration of the noble gas, krypton, has been achieved, the 5*s* orbital is more stable than the 4*d* orbitals. The next two electrons therefore enter the 5*s* orbital, but then the 4*d* orbitals fall to lower energy than the 5*p* orbitals, and the second regular transition series commences with the element yttrium. Electrons continue to be added to the 4*d* orbitals until those orbitals are entirely filled at the position of the element cadmium, which has an electron configuration [Kr] $4d^{10} 5s^2$. The next six electrons enter the 5*p* orbitals until another noble gas configuration is attained at the element xenon. Analogously to the two preceding periods, the next two electrons are added to the next available orbital, namely, the 6*s* orbital, producing the next two elements, cesium and barium. At this point, however, the ordering of orbitals becomes more complex than it previously had been, because there are now unfilled 4*f* orbitals as well as the 5*d* orbitals, and the two sets have approximately the same energy. In the next element, lanthanum (atomic number 57), an electron is added to the 5*d* orbitals, but the immediately following element, cerium (atomic number 58), has two electrons in the 4*f* orbitals and none in the 5*d* orbitals. Through the next 12 elements the additional electrons enter the 4*f* orbitals, although the 5*d* orbitals are of only slightly higher energy. This set of elements, spanning the range from lanthanum, where the 4*f* orbitals were still vacant or about to be filled, through lutetium, in which the 4*f* orbitals are completely filled by 14 electrons, makes up the lanthanides, mentioned above.

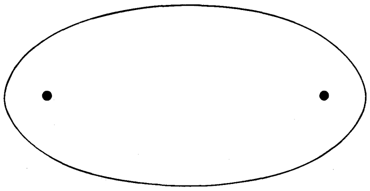
Second transition series

At this point the next available orbitals are the 5*d* orbitals, and the elements hafnium through gold, the third regular transition series, correspond to the successive filling of these 5*d* orbitals. Following this series there are again *p* orbitals (6*p*) to be filled, and when this is accomplished the noble gas radon is reached.

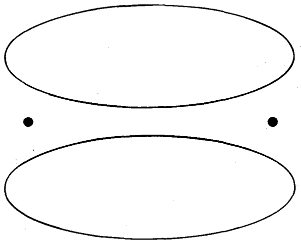
Third transition series

Molecular orbitals. If two atoms are close together, some of their orbitals may overlap and participate in the formation of molecular orbitals. Electrons that occupy a molecular orbital interact with the nuclei of both atoms: if this interaction results in a total energy less than that of the separated atoms, as is the case if the orbital lies mainly in the region between the two nuclei, the orbital is said to be a bonding orbital and its occupancy by electrons constitutes a covalent bond that links the atoms together in compound formation and in which the electrons are said to be shared. If the occupation of an orbital by electrons raises the energy of the system, as is the case if the orbital lies mainly outside the region between the two nuclei, that orbital is said to be antibonding; the presence of electrons in such orbitals tends to offset the attractive force derived from the bonding electrons.

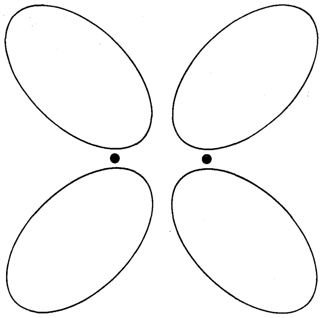
A bonding or an antibonding molecular orbital may be disposed along the line passing through the two nuclei (Figure 13A), in which case it is designated by the Greek letter σ (sigma); or it may occupy regions approximately parallel to that line (Figure 13B and 13C) and be designated π (pi).



A bonding σ orbital



B bonding π orbital



C antibonding π orbital

Figure 13: Electron distributions in molecular orbitals (see text).

DISCOVERY OF THE TRANSITION ELEMENTS

The most abundant transition element in the Earth's solid crust is iron, which is fourth among all elements and second (to aluminum) among metals in crustal abundance. The elements titanium, manganese, zirconium, vanadium, and chromium also have abundances in excess of 100 grams (3.5 ounces) per ton. Some of the most important and useful transition elements have very low crustal abundances—*e.g.,* tungsten, platinum, gold, and silver. Table 23 lists the abundances of the transition elements.

Four of the regular transition elements were known to the ancients: iron (*ferrum*), copper (*cuprum*), silver (*argentum*), and gold (*aurum*). Their chemical symbols (Fe, Cu, Ag, Au), in fact, are derived from their alchemical (Latin) names rather than their contemporary names. The other regular transition elements were discovered (or recognized as elements) after the early 18th century. The transition element most recently discovered in nature is rhenium (atomic number 75), which in 1925 was detected in platinum ores and in the niobium mineral columbite.

Technetium (atomic number 43) is the only *d*-block element that has not been isolated from the Earth's crust. All isotopes of technetium are radioactive; the half-life of even the stablest isotope, technetium-97, is too short to permit the survival of primordial technetium in the Earth's crust, and claims that it has been isolated or detected there must be considered erroneous. Technetium can be isolated in considerable quantities, however, from the fission products of nuclear reactors, and it is at least as readily available for chemical study as the naturally occurring similar element rhenium, of which there are no concentrated ores.

TRANSITION-METAL CATALYSTS

Types of transition-metal catalysts

One important use of transition metals and their compounds is as catalysts for a variety of industrial processes, mostly in the petroleum and polymer (plastics, fibres) industries, in which organic molecules are isomerized, built up from simple molecules, oxidized, hydrogenated, or caused to polymerize. Only a few of the most important such processes and their catalysts can be mentioned here. Catalysts are of two physical types: homogeneous (*i.e.,* dissolved in the reaction mixture) and heterogeneous (*i.e.,* constituting a solid phase separate from and insoluble in the reaction mixture). Both types are represented on the industrial scene, but the latter are much more common.

The introduction of catalysts that allow polymerization to be carried out at relatively low temperatures and pressures revolutionized the production of polyethylene and polypropylene. Previously polyethylene had to be made by a process requiring pressures of about 1,000 atmospheres, and polypropylene of useful properties was not commercially important. The catalysts devised and applied during the early 1950s are prepared from titanium tetrachloride and an aluminum alkyl such as triethylaluminum: the precipitated, titanium-containing solid, plus the excess aluminum alkyl in solution, constitutes the catalyst. A very different sort of catalyst, consisting of chromium(VI) oxide dispersed on silica-alumina, performs similarly in polymerizing ethylene but cannot produce a useful form of polypropylene.

Chromium in the form of chromium(III) oxide on alumina is the major industrial catalyst for transforming saturated hydrocarbons (*i.e.,* those in which all available valence bonds of the atoms are attached to other atoms) to useful olefins (unsaturated organic compounds), chiefly *n*-butane to butylene and butadiene.

Ammonia synthesis

Iron-containing catalysts are used in various processes of which the most notable is that for producing ammonia from nitrogen and hydrogen. This process, developed early in the 20th century, represents the first major industrial application of transition metal catalysis. The catalyst is magnetic triiron tetroxide (Fe_3O_4), "promoted" by the addition of small quantities of potassium oxide, aluminum oxide, calcium oxide, and silica.

Molybdenum in molybdenum(VI) oxide–aluminum oxide mixtures or in cobalt oxide–molybdenum oxide–aluminum oxide mixtures finds many applications, such as desulfurizing gases and liquids; removing lead, nickel, and vanadium from petroleum refinery feed stocks; and in modifying, or "reforming," crude petroleum fractions to increase their octane rating.

Olefins that are free of such impurities as carbon monoxide, sulfur, halogen, and compounds of arsenic or lead (catalyst poisons), can be hydrogenated (*i.e.,* combined with hydrogen) at atmospheric pressure and room temperature, using various types of active nickel preparations as catalysts.

Copper is a component of a variety of catalysts, of which the copper chromites, used for selective hydrogenation of carbonyl groups (compounds containing the divalent organic radical $>$CO), are perhaps the most important. Fats and oils can also be hydrogenated to alcohols using copper catalysts. Palladium chloride together with copper(I) salts forms a homogeneous catalyst for oxidizing ethylene to acetaldehyde (Wacker process), but commercial use of this process has been abandoned. Metallic platinum has a broad spectrum of catalytic activities. One of the most important in terms of tonnage production is in catalytic reforming of petroleum fractions to improve antiknock quality of gasoline. Silver oxide, on an inactive, refractory support, catalyzes oxidation of ethylene to ethylene oxide.

The oxo process (also called hydroformylation), in which certain olefins react with carbon monoxide and hydrogen to give aldehydes (a class of organic compounds that contain the group —CHO), requires a homogeneous catalyst containing a transition metal. The first practicable one was hydrogen tetracarbonylcobaltate(-I), $HCo(CO)_4$, which is formed in the reaction mixture by action of hydrogen on octacarbonyldicobalt, $Co_2(CO)_8$. More recently rhodium complexes have been found to have greater activity at lower temperatures and pressures and to be more easily recovered. The net reaction in the oxo process is represented by

$$\underset{\text{a terminal olefin}}{RCH{=}CH_2} + \underset{\text{carbon monoxide}}{CO} + \underset{\text{hydrogen}}{H_2} \longrightarrow \underset{\text{an aldehyde}}{RCH_2CH_2CHO.}$$

BIOLOGICAL FUNCTIONS OF TRANSITION METALS

Several transition elements are important to the chemistry of living systems, the most familiar examples being iron, cobalt, copper, and molybdenum. Iron is by far the most widespread and important transition metal that has a function in living systems; proteins containing iron participate in two main processes, oxygen transport and electron transfer (*i.e.,* oxidation–reduction) reactions. There are also a number of substances that act to store and transport iron itself.

Vitamin B_{12}

Though cobalt is understood to be an essential trace element in animal nutrition, the only detailed chemical knowledge of its biochemical action has to do with vitamin B_{12} and related co-enzymes. These molecules contain one atom of cobalt bound in a macrocyclic ring (*i.e.,* one consisting of many atoms) called corrin, which is similar to a porphyrin ring. Copper is found in both plants and animals, and numerous copper-containing proteins have been isolated. The blood of many lower animals, such as mollusks, cephalopods, gastropods, and decapods, contains respiratory proteins called hemocyanins, which contain copper atoms (but no heme) and appear to bind one oxygen molecule per two copper atoms. Human serum contains a glycoprotein called ceruloplasmin, the molecule of which contains eight copper atoms; its biological function is still uncertain. Other proteins, called cerebrocuprein, erythrocuprein, and hepatocuprein, that are found in the mammalian brain, erythrocytes, and liver, respectively, contain about 60 percent of the total copper in those tissues; their functions are still unknown. There are a number of copper-containing enzymes; examples are (1) ascorbic acid oxidase (an oxidase is an oxidizing enzyme), which contains eight atoms of copper per molecule; it is widely distributed in plants and microorganisms; (2) cytochrome oxidase, which contains heme and copper in a 1 : 1 ratio; (3) tyrosinases, which catalyze the formation of melanin (brownish-black pigments occurring in hair, skin, and retina of higher animals) and were the first enzymes in which copper was shown to be essential to function.

Nitrogen fixation

Vanadium occurs widely in petroleum, notably that from Venezuela, and can be isolated as porphyrin complexes, the origin of which is not known. Vanadium is present in high concentrations in blood cells (vanadocytes) of certain ascidians (sea squirts), apparently in a curious, complex, and poorly understood protein-containing substance called hemovanadin, thought to serve in oxygen transport. Molybdenum is believed to be a necessary trace element in animal diets, but its function and the minimum levels have not been established. Nitrogen-fixing bacteria utilize enzymes that contain both molybdenum and iron. One such enzyme, or at least a part of it that has been isolated in the crystalline state, contains two atoms of molybdenum and 40 atoms of iron. This protein in association with another, which contains only iron, can catalyze the reduction of nitrogen gas to nitrogen compounds.

THEORIES OF TRANSITION-METAL COMPLEXES

As has been noted, partially filled *d* orbitals account for the characteristic chemical properties of the regular transition elements, both as a class and as individuals. The interpretation and understanding of the chemical and physical properties of these elements thus depends heavily upon the description of these d^n (*n* is one or more but fewer than ten) electron configurations. The five orbitals of each *d* shell, regardless of principal quantum number, have the shapes and designations shown in Figure 14. The radial extent or size changes with principal quantum number, but the shapes are characteristic for all sets.

d-orbital subsets

For an atom or ion in free space and not subject to any electromagnetic field, the five orbitals of a set have equal energies; but in chemical compounds, the surrounding atoms cause the set to separate into subsets, differing in energy. This separation has a profound influence upon the spectroscopic properties (including colour) and magnetic properties of the substance, and the analysis of such effects has stimulated a great deal of theoretical and experimental effort. The most common situations to be considered are those in which a transition-metal ion occurs in an octahedron or a tetrahedron of surrounding anions or other ligands. Such cases illustrate the handling of the problem of the electronic structure of transition-metal ions in compounds.

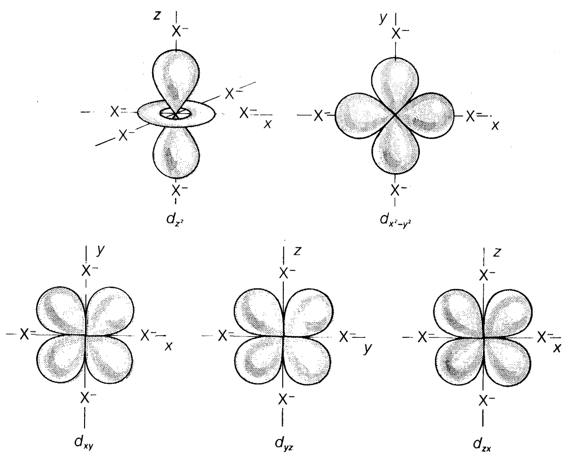


Figure 14: The shapes of the *d* orbitals and their conventional designations. The symbols X^- represent some or all of a set of six octahedrally located anions.

Three different theoretical approaches have been used: (1) the valence-bond treatment, pioneered in the United States by Linus Pauling; (2) the crystal-field theory or its more sophisticated form, the ligand-field theory, first proposed by Hans Bethe and developed extensively by the U.S. physicist J.H. Van Vleck; and (3) the molecular orbital theory, the application of which to transition-metal complexes was first discussed by Van Vleck. The second and third methods are used almost exclusively, and only those two will be outlined here.

Crystal-field and ligand-field theories. The crystal-field theory (CFT) employs an extreme electrostatic model, in which the ligands are treated as point charges (*i.e.,* as if the entire negative charge were concentrated at a single point in space) if they are anions or as point dipoles (*i.e.,* pairs of particles having equal and opposite charges that are separated by a finite distance) if they are neutral molecules. These extreme approximations are useful because they preserve certain essential features of the actual physical condition and yet reduce the mathematical problem to one that can be solved. Since the CFT treatment makes no allowance for covalence in the metal–ligand bonds, it necessarily does not account for all aspects of the electronic structures of complexes, even the most ionic ones. It can be amended, however, by empirical adjustment of certain parameters, to allow for some of the effects of covalence, without sacrificing mathematical convenience. This version of the theory, which can correlate much of the data on complexes of metals in their normal oxidation states, is called ligand-field theory (LFT).

Octahedral complexes

Orbital splitting patterns. Crystal-field theory treats the question of how the energies of a set of *d* electrons or orbitals are split when a set of ligands is placed around a central metal ion; it does so by treating the ligands as a set of negative charges. The simplest case to consider is that of an ion with a single *d* electron, surrounded by six negative charges at the vertices of an octahedron. This arrangement is defined, relative to a set of Cartesian axes, *x, y, z,* shown in Figure 15. By comparing the shapes of the *d* orbitals (see Figure 14) with this arrangement, it can be seen that the d_{xy}, d_{xz}, and d_{yz} orbitals have equivalent relationships to the set of charges. Thus, the electron will be repelled to the same extent by the negative charges regardless of which of the three orbitals it occupies. The

Tetrahedral complexes

three orbitals thus have equal energy and are called triply degenerate. It is not particularly obvious from a pictorial argument, but mathematical analysis shows that each of the other two orbitals, d_{z^2} and $d_{x^2-y^2}$, causes the electron to experience the same amount of electrostatic repulsion from the surrounding charges; they are described as doubly degenerate. Thus, the first important conclusion of the crystal-field theory is that the spatial relationships of the d orbitals to the surrounding charges cause the set of five d orbitals to be split into two subsets; the orbitals within each subset are equivalent to each other and are thus degenerate, but they are no longer degenerate with those in the other subset. Those in the subset consisting of the d_{xy}, d_{yz}, and d_{zx} orbitals are known as the t_{2g} orbitals; the d_{z^2} and $d_{x^2-y^2}$, pair are called the e_g orbitals. When the surrounding charges are located at the vertices of an octahedron, the two e_g orbitals are of higher energy than the three t_{2g} orbitals. When an ion is surrounded by a tetrahedrally arranged set of four negative charges, the d orbitals also split into a set of three and a set of two, but the energy order is the reverse of that in the octahedral case. Beginning with these splitting patterns it is possible to elaborate a detailed account of the magnetic and spectroscopic properties of transition-metal ions in their compounds.

Factors influencing magnetic properties

Magnetic and spectroscopic properties. Discussion of magnetic properties must begin with the basic question of how many unpaired electrons will be present. This is decided by the competition between two factors: (1) the electrons tend to occupy separate orbitals since this minimizes repulsions between them; and (2) the electrons also tend to occupy orbitals of lower energy in preference to higher ones. For ions with one, two, and three d electrons in octahedral fields, the first factor dictates, without opposition from the second, that the electrons occupy separate t_{2g} orbitals; such ions therefore always have 1, 2, and 3 unpaired electrons. Similarly, for configurations of eight and nine d electrons, the only possibilities are six electrons in t_{2g} orbitals and two in e_g orbitals, and six electrons in t_{2g} orbitals and three in e_g orbitals ($t_{2g}^6e_g^2$ and $t_{2g}^6e_g^3$), and such ions must always have 2 and 1 unpaired electrons, respectively. For cases of d^4, d^5, d^6, and d^7 configurations, the number of unpaired electrons depends on which of the two factors dominates. If the splitting (the energy difference between e_g and t_{2g} orbitals) is relatively small, control by the first factor produces the high-spin configurations, $t_{2g}^3e_g$, $t_{2g}^3e_g^2$, $t_{2g}^4e_g^2$, and $t_{2g}^5e_g^2$, which have 4, 5, 4, and 3 unpaired electrons, respectively. When the difference is relatively large, control by the second factor produces the low-spin configurations t_{2g}^4, t_{2g}^5, t_{2g}^6, and $t_{2g}^6e_g$, with 2, 1, 0, and 1 unpaired electrons, respectively. In tetrahedral complexes the splitting is always sufficiently small so that the first factor dominates and high-spin states are always obtained. Once the number of unpaired eletrons is known, numerical values of the magnetic moments are calculated by taking account of the interaction of the orbital-angular momentum with the spin-angular momentum.

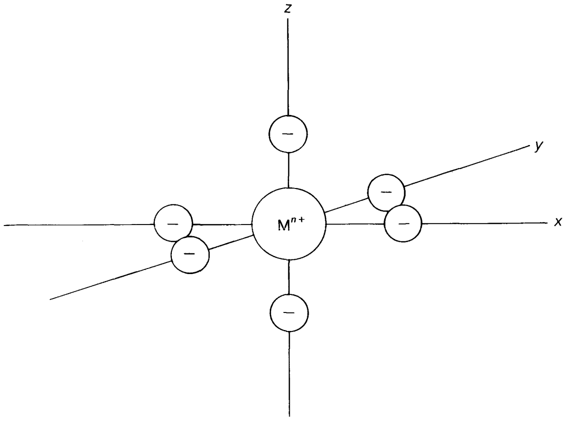


Figure 15: An octahedrally arranged set of negative charges in relation to a set of Cartesian coordinates with a positively charged metal ion (M^{n+} at the centre).

The spectra emitted by the electronic structures of transition-metal ions, which are responsible for their colours, can be understood in terms of the CFT splitting patterns. For a d^1 ion in an octahedral field, a single electronic transition, $t_{2g} \rightarrow e_g$, is expected; that is, the absorption of light raises the energy of an electron and causes it to pass from the low-energy t_{2g} orbital to the high-energy e_g orbital. The titanium(III), Ti^{3+}, ion, for example, has just a single absorption band in the visible region of the spectrum, at a wavelength of about 5,000 nanometres, corresponding to an energy of 20,000 cm^{-1}, which is assigned to this transition. This says that the orbital energy difference for Ti^{3+} in $[Ti(H_2O)_6]^{3+}$ is 20,000 per centimetre. For configurations with more than one d electron, electronic interactions require that an elaborate treatment, which cannot be explained here, be used.

The difference in the colours of hexaquonickel(II) ion, $[Ni(H_2O)_6]^{2+}$ (green), and tris (ethylenediamine) nickel(II) ion, $[Ni(H_2NCH_2CH_2NH_2)_3]^{2+}$ (purple), reflects the fact that the six nitrogen atoms cause a greater splitting than the six oxygen atoms.

In general, the relative magnitudes of d orbital splittings for a given ion with different ligand sets fall in a consistent order. This ordering of ligands according to their ability to split the d orbitals was discovered empirically in the 1930s and is called the spectrochemical series. A short list of ligands in order of their splitting power is fluoride (weakest), water, thiocyanate, pyridine or ammonia, ethylenediamine, nitrite, cyanide (strongest).

Jahn-Teller effect. According to the Jahn-Teller theorem, any molecule or complex ion in an electronically degenerate state will be unstable relative to a configuration of lower symmetry in which the degeneracy is absent. The chief applications of this theorem in transition-metal chemistry are in connection with octahedrally coordinated metal ions with high-spin d^4, low-spin d^7, and d^9 configurations; in each of these cases the t_{2g} orbitals are all equally occupied (either all half filled or all filled) and there is a single electron or a single vacancy in the e_g orbitals. Either an e_g or an e_g^3 configuration gives rise to a doubly degenerate (E) ground state, and thus a distortion of the octahedron is expected. In other words, high-spin d^4, low-spin d^7, and d^9 ions should be found in distorted, not regular, octahedral environments. It has been found by experiment that, with few possible exceptions, this is the case. The exceptions do not necessarily constitute violations of the Jahn-Teller theorem, because the slightness of the distortions may be within experimental error.

Distorted complexes

By far the most common form of distortion in the three cases given above is an elongation of the octahedron along one of its fourfold axes. Thus, two opposite ligands are farther from the metal ion than the other, coplanar (*i.e.,* lying in one plane) set of four. Such effects have been observed in compounds and complexes of high-spin chromium(II) and manganese(III) ions, ($t_{2g}^3e_g$); low-spin cobalt(II) and nickel(III) ions, ($t_{2g}^6e_g$); and copper(II) and silver(I) ions ($t_{2g}^6e_g^3$). The greatest body of data available is for copper(II) compounds, and some representative sets of copper–ligand distances are copper(II) chloride, four chloride ions at 2.30 Å and two at 2.95 Å; copper(II) bromide, four bromide ions at 2.40 Å and two at 3.18 Å.

Molecular-orbital theory. The molecular-orbital (MO) treatment of the electronic structures of transition-metal complexes is, in principle, a more flexible approach than the CFT or LFT treatments. Because a great many complexes and compounds in the ordinary oxidation states (II, III) of the transition metals are substantially ionic, the CFT and LFT treatments are useful though not exact. In the MO method, the ligand-atom orbitals and all of the valence-shell orbitals of the metal atom are presumed to interact to form molecular orbitals.

The interaction of a metal ion from the first transition series and an octachedral set of ligands that interact only through orbitals directed toward the metal ion results in the separation of the d orbitals of the metal ion into two sets of molecular orbitals; a t_{2g} set and a higher energy e_g set. This is the same pattern as that obtained in the

CFT treatment, but the difference is that now the t_{2g} and e_g orbitals are regarded not as pure d orbitals but as molecular orbitals, having both metal d orbital and ligand orbital components. Thus, the modifications of CFT that lead to LFT—that is, inclusion of covalence effects—are in harmony with the basic ideas of the MO treatment.

THE ELEMENTS OF THE FIRST TRANSITION SERIES

General chemical trends of the first transition series

Although the transition elements have many general chemical similarities, each one has a detailed chemistry of its own. The closest relationships are usually to be found among the three elements in each vertical group in the periodic table, although within each group the element of the first series usually differs more from the other two than they differ from each other. Most of the first series elements are more familiar and technically important than the heavier members of their vertical group. Thus, it is convenient to discuss first each of the metals of the first transition series and then to discuss the remaining ones in pairs.

Before considering each of the individual elements, a few of the chemical trends to be found in the first transition series may be capsulized.

1. From titanium to manganese the highest valence exhibited, which usually is found only in oxo compounds, fluorides, or chlorides, corresponds to the total number of $3d$ and $4s$ electrons in the atom. The stability of this highest oxidation state decreases from titanium(IV) to manganese(VII). Following manganese—that is, for iron, cobalt, and nickel—oxidation states corresponding to the loss of all $3d$ and $4s$ electrons do not occur; higher oxidation states in general become progressively more difficult to attain because the increasing nuclear charge causes the $3d$ electrons to be more tightly bound. Very high oxidation states occur only for chromium(V, VI), manganese(V, VI, VII), and iron(V, VI) and apart from the fluorides, such as chromium(V) fluoride, CrF_5, and chromium (VI) fluoride, CrF_6, and oxofluorides such as manganese(VII) trioxide fluoride, MnO_3F, the main chemistry in these oxidation states is that of oxo anions such as manganate(VII), MnO_4^-; chromate(VI), CrO_4^{2-}; and ferrate(VI), FeO_4^{2-}. All of these compounds are powerful oxidizing agents.
2. The oxides of each element become more acidic with increasing oxidation number, and the halides become more covalent and susceptible to hydrolysis.
3. In the oxo anions characteristic of the higher valence states the metal atom is tetrahedrally surrounded by oxygen atoms, whereas in the oxides formed in the lower valence states the atoms are usually octahedrally coordinated.
4. In the oxidation states II and III, complexes in aqueous solution or in crystals are usually four-, five- or six-coordinated.
5. Oxidation states lower than II are not found in the ordinary chemistries of the transition elements, except for copper. The lower oxidation states are, however, attainable for all the elements using ligands of the carbon monoxide type.

Titanium. The preparation of pure titanium is difficult because of its reactivity. Titanium cannot be obtained by the common method of reducing the oxide with carbon because a very stable carbide is readily produced and, moreover, the metal is quite reactive toward oxygen and nitrogen at elevated temperatures. Therefore, special processes have been devised. One of the ores, such as ilmenite ($FeTiO_3$) or rutile (TiO_2), is treated at red heat with carbon and chlorine to yield titanium tetrachloride, $TiCl_4$, which is fractionally distilled to eliminate impurities such as ferric chloride, $FeCl_3$. The $TiCl_4$ is then reduced with molten magnesium at about 800° C in an atmosphere of argon, and metallic titanium is produced as a spongy mass from which the excess of magnesium and magnesium chloride can be removed by volatilization at about 1,000° C. The sponge may then be fused in an atmosphere of argon or helium in an electric arc and be cast into ingots. On the laboratory scale, extremely pure titanium can be made by vaporizing the tetraiodide, TiI_4, in very pure form and decomposing it on a hot wire in vacuum.

itanium
roduction

The atoms are arranged in a hexagonal, close-packed lattice in titanium, which resembles other transition metals such as iron and nickel in being hard, refractory, and a good conductor of heat and electricity. It is quite light in comparison to other metals of similar mechanical and thermal properties, however, as well as unusually resistant to certain types of corrosion, and therefore it has come into demand for special applications in engineering.

Although titanium is rather unreactive at ordinary temperatures, it combines directly with many nonmetals, such as hydrogen, the halogens, oxygen, nitrogen, carbon, boron, silicon, and sulfur at elevated temperatures. The resulting nitride (TiN), carbide (TiC), and borides (TiB and TiB_2) are interstitial compounds that are very stable, hard, and refractory. Titanium is not attacked by mineral acids at room temperature or by hot aqueous alkali; it dissolves in hot hydrochloric acid, giving trivalent titanium species, and hot nitric acid converts it into a hydrous oxide that is rather insoluble in acid or base. The best solvents for the metal are hydrofluoric acid or other acids to which fluoride ions have been added; such mediums dissolve titanium and hold it in solution because of the formation of fluoro complexes.

Preparation and uses of vanadium

Vanadium. The preparation of very pure vanadium is difficult because the metal is quite reactive toward oxygen, nitrogen, and carbon at elevated temperatures. Because the major commercial use of vanadium is in steel and cast iron, to which it lends ductility and shock resistance, vanadium is produced commercially mainly in the form of an iron alloy called ferrovanadium. When the very pure metal is required, it may be obtained by processes described in the foregoing discussion of titanium. Very pure vanadium metal resembles titanium in being quite corrosion resistant, hard, and steel gray in colour. In the massive state it is not attacked by air, water, alkalies, or nonoxidizing acids other than hydrofluoric acid. It dissolves in nitric acid, concentrated sulfuric acid, or aqua regia and at high temperatures will combine with nearly all nonmetals. For vanadium the important oxidation states are II, III, IV, and V.

Chromium. The name of the element chromium (from Greek *chrōmos,* "colour") connotes the pronounced and varied colorations of chromium compounds. The red colour of ruby and the green colour of emerald, to mention only gems, are due to small amounts of chromium. The free metal is never found in nature; most ores consist of the mineral chromite, the ideal formula of which is $FeCr_2O_4$. Natural deposits are usually contaminated with magnesium, aluminum, and silica and their chromium content varies from 42 to 56 percent. One of the chief uses of chromium is in ferrous alloys, for which the pure metal is not required. Accordingly, chromite is often reduced with carbon in a furnace, producing the alloy ferrochromium, which contains iron and chromium in an atom ratio of approximately 1 to 2. To obtain pure chromium, chromite is first treated with molten alkali and oxygen, converting all of the chromium to the alkali chromate, and the latter is dissolved in water and eventually precipitated as sodium dichromate, $Na_2Cr_2O_7$. The dichromate is then reduced with carbon to chromium(III) oxide, Cr_2O_3, and that oxide in turn reduced with aluminum to give the chromium metal.

The metal is white, hard, lustrous, and brittle and is extremely resistant to ordinary corrosive reagents; this resistance accounts for its extensive use as an electroplated protective coating. Chromium metal dissolves rather readily in nonoxidizing mineral acids—for example, hydrochloric or sulfuric acids; it does not dissolve in cold aqua regia or in nitric acid because of a curious phenomenon known as passivation, in which the metal presumably becomes covered with a thin unreactive coating. At elevated temperatures chromium unites directly with the halogens or with sulfur, silicon, boron, nitrogen, carbon, or oxygen.

Manganese. As in the case of titanium, vanadium, and chromium, the highest oxidation state (+7) of manganese corresponds to the total number of $3d$ and $4s$ electrons. That state occurs only in the oxo species manganate (VII) (MnO_4^-), manganese(VII) oxide (Mn_2O_7), and manganese(VII) trioxide fluoride (MnO_3F), which show some similarity to corresponding compounds of the halogens—for example, in the instability of the oxide. Manganese in oxidation state +7 is powerfully oxidizing, usually being reduced to divalent manganese. The intermediate oxidation states are known, but, except for some compounds in the +3 and +4 states, they are not particularly important.

Manganese, first recognized as an element by the Swedish chemist Carl Wilhelm Scheele in 1774, is second only to iron among the transition elements in its abundance in the Earth's crust and is roughly similar to iron in its physical and chemical properties but is harder and more brittle. Although it is widely distributed, manganese occurs in a number of substantial deposits, mainly the oxides, of which the mineral pyrolusite (manganese(IV) oxide, MnO_2) is the most important. Manganese is quite electropositive, dissolving very readily in dilute non-oxidizing acids. Although relatively unreactive toward nonmetals at room temperature, it reacts with many at elevated temperatures. Thus, manganese burns in chlorine to give manganese(II) chloride, $MnCl_2$; reacts with flourine to give manganese(II) and (III) fluorides, MnF_2 and MnF_3; and burns in nitrogen at about 1,200° C to give manganese (II) nitride, Mn_3N_2, and in oxygen to give trimanganese tetroxide Mn_3O_4. Manganese also combines directly with boron, carbon, sulfur, silicon, or phosphorus, but not with hydrogen. The most important oxidation states for manganese are +2, +3, +7.

Iron. Hundreds of mineral species contain iron as a constituent, and igneous rocks average about 5 percent iron content. The major iron ores are the minerals hematite [iron(III) oxide, Fe_2O_3], magnetite [triiron tetroxide, Fe_3O_4], limonite [iron(III) oxide hydroxide, FeO(OH)], and siderite [iron(II) carbonate, $FeCO_3$].

For the element iron the trends already noted in the relative stabilities of oxidation states among elements of the first transition series are continued, except that there is no compound or chemically important circumstance in which the oxidation state of iron is equal to the total number of its valence shell electrons, eight; the highest known oxidation state is +6, which is rare and unimportant. Even the trivalent state, which is important at the position of chromium in the periodic table, loses ground to the divalent state at the position of iron. This trend continues in the remaining elements with the sole exception of trivalent cobalt, which is stable in a host of complexes.

The Mössbauer effect

Iron compounds are amenable to study by taking advantage of a phenomenon known as the Mössbauer effect (the phenomenon of a gamma ray being absorbed and reradiated by a nucleus without recoil). Although the Mössbauer effect has been observed for about one-third of the elements, it is particularly for iron (and to a lesser extent tin) that the effect has been a major research tool for the chemist. In the case of iron the effect depends on the fact that the nucleus of iron-57 can be excited to a high energy state by the absorption of gamma radiation of very sharply defined frequency that is influenced by the oxidation state, electron configuration, and chemical environment of the iron atom and can thus be used as a probe of its chemical behaviour.

Pure iron is a white, lustrous metal that is not particularly hard and is quite reactive. As commonly available, however, iron almost always contains small amounts of carbon, which modify its properties greatly. In a very finely divided state metallic iron is pyrophoric (*i.e.,* it ignites spontaneously). It combines vigorously with chlorine on mild heating and also with a variety of other nonmetals, including all of the halogens, sulfur, phosphorus, boron, carbon, and silicon (the carbide and silicide phases play major roles in the technical metallurgy of iron). Metallic iron dissolves readily in dilute mineral acids. With nonoxidizing acids and in the absence of air divalent iron is obtained; with air present or when warm dilute nitric acid is used, some of the iron goes into solution as the iron(III) ion, Fe^{3+}. Very strongly oxidizing mediums, for example, concentrated nitric acid or acids containing dichromate, passivate iron (*i.e.,* cause it to lose its normal chemical activity), however, much as they do chromium (see above). Air-free water and dilute air-free hydroxides have little effect on the metal, but it is attacked by hot concentrated sodium hydroxide.

Rusting of iron

Under normal atmospheric conditions iron rusts to form hydrated oxides, a process that occurs in two steps: first, iron dissolves in the acid solution produced by the moisture and the carbon dioxide of the air, to form ferrous iron—*i.e.,* iron(II)—and liberate hydrogen; second, oxygen from the air oxidizes the ferrous iron to form hydrated iron(III) oxide.

The stable structure of iron up to 766° C is that of a crystal form called body-centred cubic, and it is ferromagnetic (*i.e.,* capable of being permanently magnetized). Above this temperature the structure is retained but the magnetic properties change to simple paramagnetic (*i.e.,* capable of being only weakly magnetized and only as long as the magnetizing field is present). At 906° C the structure changes to a crystal form called cubic close-packed; but there is another structural change at 1,401° C, and iron reverts to the body-centred cubic structure.

Cobalt. The trends toward decreased stability for very high oxidation states and increased stability of the +2 state relative to the +3 state, which have been evident through the series titanium, vanadium, chromium, manganese, iron, persist with cobalt. In fact there is a complete absence of oxidation states higher than +4 under chemically significant conditions, and even the latter may be represented by only a few incompletely characterized compounds. The +3 state is relatively unstable in simple compounds, but forms a host of low-spin complexes, most of them involving ammonia or amines as ligands. The oxidation state +1 is perhaps better known for cobalt than for any other element in the first transition series except copper.

The abundance of cobalt in the Earth's crust is relatively low, and the element is widely distributed; even the best cobalt ores contain only low concentrations of the element and require complex processing to concentrate and extract it. In general, little cobalt ore is mined for the cobalt content: the metal is recovered as a by-product from the ores of other metals, particularly copper and nickel.

Cobalt is a hard, bluish-white metal and, along with iron and nickel, is one of the three metals that are ferromagnetic at room temperature. It dissolves slowly in dilute mineral acids, does not combine directly with either hydrogen or nitrogen, but will combine, on heating, with carbon, phosphorus, or sulfur. Cobalt is also attacked by oxygen and by water vapour at elevated temperatures, with the result that cobalt(II) oxide, CoO, is produced.

Nickel. Typical compounds of nickel in nature, in which it occurs primarily as minerals in combination with arsenic, antimony, and sulfur, are nickel(II) sulfide, NiS; nickel (III) arsenide, NiAs; nickel(III) antimonide, NiSb; nickel (III) diarsenide, $NiAs_2$; nickel(III) thioarsenide, NiAsS; and nickel(III) thioantimonide, NiSbS. Some commercially important deposits contain the mineral garnierite, a magnesium–nickel silicate of variable composition. Also certain varieties of the iron mineral pyrrhotite contain 3 to 5 percent nickel. Elemental nickel is found alloyed with iron in many meteors, and the central regions of the Earth are believed to contain considerable quantities. The free metal is never found in the Earth's crust. The metallurgy of nickel is complicated in its details, many of which vary widely, according to the particular ore being processed. In general the ore is transformed to nickel(III) sulfide, Ni_2S_3, which is roasted in air to give nickel(II) oxide, NiO, which is then reduced with carbon to obtain the metal. Some high-purity nickel is made by the carbonyl process mentioned earlier.

Properties and reactions of nickel

Nickel is a silver-white metal with high electrical and thermal conductivity and it can be fabricated readily by the use of standard hot and cold working methods. The metal is fairly resistant to attack by air or water at ordinary temperatures when compact and is therefore often electroplated as a protective coating. Nickel reacts only slowly with fluorine, eventually developing a protective coating of the fluoride, and therefore is used as the pure metal or in the form of alloys such as Monel metal (which contains some 60 to 70 percent nickel, 30 to 40 percent copper, and small amounts of other metals such as iron) in equipment for handling fluorine gas and corrosive fluorides. Nickel is ferromagnetic at ordinary temperatures, although not as strongly so as iron, and is less electropositive than iron but dissolves readily in dilute mineral acids.

Copper. Copper is distributed widely in nature in the form of the free metal, sulfides, arsenides, chlorides, and carbonates. It is extracted by oxidative roasting and smelting, followed by electrodeposition from sulfate solutions, and it is technologically important, particularly in alloys, as an electroplated coating and as an exceptionally good

Properties and reactions of copper

conductor of heat and electricity. The pure metal is soft, tough, ductile, and characteristically reddish in colour and is second only to silver in thermal and electrical conductivity. Metallic copper is only superficially oxidized in air, sometimes acquiring a green coating that is a mixture of hydroxo-carbonate, hydroxosulfate, and small amounts of other compounds.

Copper is a moderately noble metal, being unaffected by nonoxidizing or noncomplexing dilute acids in the absence of air; it will, however, dissolve readily in nitric acid and in sulfuric acid in the presence of oxygen and is also soluble in aqueous ammonia or potassium cyanide in presence of oxygen, because of the formation of very stable cyano complexes upon dissolution. The metal will react at red heat with oxygen to give copper(II) oxide, CuO, and, at higher temperatures, copper(I) oxide, Cu_2O. It reacts on heating with sulfur to give copper(I) sulfide, Cu_2S.

Copper forms compounds in the oxidation states +1 and +2 in its normal chemistry, although under special circumstances some compounds of trivalent copper can be prepared.

THE ELEMENTS OF THE SECOND AND THIRD TRANSITION SERIES

Differences in properties of elements in the second and third series from those of the first

While the elements in the second and third transition series for a given group have chemical properties similar to those of the element in the first series, they nevertheless show definite differences from the lighter element of the group. The following examples illustrate this point: (1) Although cobalt (of the first series) forms a considerable number of tetrahedral and octahedral complexes in its divalent state, and that valence state is characteristic in ordinary aqueous chemistry, divalent states of rhodium (second series) and iridium (third series) are rare and relatively unimportant. (2) The manganese(II) ion (Mn^{2+}) is very stable and of principal importance in the chemistry of manganese, but for technetium and rhenium the oxidation state +2 is little more than a laboratory curiosity. (3) Trivalent chromium forms a great number of complexes, which make up one of the best known aspects of the chemistry of the element; whereas trivalent molybdenum and tungsten are not particularly stable states under any conditions and form only a few complexes. (4) The oxy anions of the first-row elements in their higher oxidation states—for example, chromate(VI) and manganate(VII)—are powerful oxidizing agents the chemistry of which is essentially restricted to that function; whereas their stoichiometric analogues, such as molybdate(VI), tungstate (VI), technetate(VII), and rhenate(VII), are quite stable and have an extensive and diverse chemistry. There are, however, some cases in which quite valid and useful analogies can be found between the chemistry of the lighter element and the two heavier elements of the group. For instance, the chemistry of trivalent rhodium complexes is, in general, quite similar to that of trivalent cobalt complexes. On the whole, however, there are differences more consistently than there are similarities between the first element and the heavier elements of each group.

For the heavier transition elements, higher oxidation states are generally more stable than is the case for the elements in the first transition series; this is true not only, as has been mentioned, for the properties of the oxo anions but for the higher halides as well. Thus the heavier elements form compounds such as ruthenium(VIII) oxide, RuO_4; tungsten(VI) chloride, WCl_6; platinum(VI) fluoride, PtF_6, etc., which have no analogues among the first-row elements, whereas the chemistry of aquo ions of lower valence states, especially +2 and +3, which is such a dominant part of the chemistry of the lighter elements, is relatively unimportant for most of the heavier ones.

One of the reasons for the generally close similarity in chemistry between the second and third transition series elements of a given group is the so-called lanthanide contraction. As already described, the series of elements known as the lanthanides comes between the second and third regular transition series and is formed by the filling of the 4*f* orbitals. The 5*d* and 6*s* orbitals, which are the valence shell orbitals for the third transition series elements, are imperfectly shielded by the 4*f* electrons from the increased nuclear charge. There is consequently, a steady contraction in the size of these orbitals through the lanthanide series of elements, with the net result that the atomic and ionic radii of hafnium, which immediately follows the lanthanide series, are almost identical to the corresponding radii of the zirconium atom, which lies just above it in Group IV. Because the zirconium and hafnium atoms have almost identical sizes as well as analogous electron configurations in all of their oxidation states, their chemistry and the properties of their compounds are exceedingly similar; indeed, it is very difficult to separate the two elements because of the great similarity in properties of their compounds. In moving along the third transition series, there is a steady but slow divergence in the properties of the second and third elements in each group so that, at the end, considerable differences exist between those of palladium and platinum and of silver and gold. The differences are not as great, however, as might have been expected had the lanthanide contraction not intervened to preclude greater disparity in the orbital sizes. Although niobium and tantalum are not quite as similar as zirconium and hafnium, the differences between them are slight, and, similarly, molybdenum and tungsten, technetium and rhenium, ruthenium and osmium, and rhodium and iridium show marked similarities in their chemistries.

Zirconium and hafnium. Zirconium occurs widely in the Earth's crust, but not in concentrated deposits. The mineral zircon, $ZrSiO_4$ (zirconium orthosilicate), which is generally found in alluvial deposits in stream beds, ocean beaches, or old lake beds, is the only commercial source of zirconium. Baddeleyite, which is essentially pure zirconium(IV) oxide, ZrO_2, is the only other important zirconium mineral. These zirconium minerals generally have a hafnium content that varies from a few tenths of 1 percent to several percent. For some purposes separation of the two elements is not important: zirconium containing about 1 percent of hafnium is as acceptable as pure zirconium. In the case of the largest single use of zirconium, however, namely, as a structural and cladding material in atomic reactors, it is essential that the zirconium be essentially free of hafnium, because the usefulness of zirconium in reactors is based on its extremely low absorption cross section for neutrons. Hafnium, on the other hand, has an exceptionally high cross section, and accordingly even slight hafnium contamination nullifies the intrinsic advantage of the zirconium. Pure hafnium in fact is also used in some atomic reactors as a control element material because of its high neutron-capture cross section.

Separation of hafnium and zirconium is generally accomplished by a liquid–liquid countercurrent-extraction procedure. In the procedure, crude zirconium(IV) chloride is dissolved in an aqueous solution of ammonium thiocyanate and methyl isobutyl ketone is passed countercurrent to the aqueous mixture, with the result that the hafnium(IV) chloride is preferentially extracted.

Either of the metals, or a mixture of the two, is produced by the same process as that described in the section *Titanium* (see above). Both the metals have extremely high melting points: zirconium 1,850° C and hafnium 2,230° C. Hafnium and zirconium are fairly resistant to acids and are best dissolved in hydrofluoric acid, in which procedure the formation of anionic fluoro complexes is important in stabilizing the solution. At normal temperatures neither metal is particularly reactive, but both become quite reactive with a variety of nonmetals at elevated temperatures.

The atomic radii of zirconium and hafnium are 1.45 and 1.44 Å, respectively, while the radii of the ions are Zr^{4+}, 0.74 Å, and Hf^{4+}, 0.75 Å. The virtual identity of atomic and ionic sizes, resulting from the lanthanide contraction, has the effect of making the chemical behaviour of these two elements more similar than for any other pair of elements known. Although the chemistry of hafnium has been studied less than that of zirconium, the two are so similar that only very small quantitative differences—for example, in solubilities and volatilities of compounds—would be expected in cases that have not actually been investigated.

The most important respect in which these two elements differ from titanium is that lower oxidation states are of minor importance; there are relatively few compounds of hafnium or zirconium in other than their tetravalent states. The increased size of the atoms makes the oxides more basic and the aqueous chemistry somewhat more extensive, and permits the attainment of coordination numbers 7 and, quite frequently, 8 in a number of zirconium and hafnium compounds.

Close association in nature

Niobium and tantalum. Although not quite as similar to each other as are hafnium and zirconium, niobium and tantalum differ but little in their chemistry, and, thus, separation problems are significant. Due to their great chemical similarity and consequent close association in nature, in fact, the establishment of the individual identities of the two elements was very difficult. Niobium was probably first discovered in 1801, in an ore sample from Connecticut, by an English chemist, Charles Hatchett, who named the element columbium in honour of the country of its origin, Columbia being a synonym for the United States. In 1844 a German chemist, Heinrich Rose, discovered what he considered to be a new element occurring along with tantalum and named it niobium after Niobe, the mythological goddess who was the daughter of Tantalus. After considerable controversy it was decided that columbium and niobium were the same element, and eventually international agreement was reached to adopt the name niobium. The name columbium is often seen, however, even today, especially in the metallurgical literature.

Niobium is roughly ten times more abundant in the crust of the Earth than is tantalum. The main commercial sources of both elements are the columbite–tantalite series of minerals, in which columbite ($FeNb_2O_6$) and tantalite ($FeTa_2O_6$) occur in highly variable ratios. Niobium is also obtainable from pyrochlore, a calcium sodium niobate. The production procedures for the metals are complex, the major problem being separation of the two elements, which is solved mainly by use of a liquid–liquid extraction process. The metals themselves are obtained by either electrolysis of fused salts or reduction of fluoro complexes with a very reactive metal such as sodium. Both metals are bright, have high melting points (niobium, 2,468° C, tantalum, 2,996° C), and are quite resistant to acids. They can best be dissolved in a mixture of nitric and hydrofluoric acids.

Molybdenum and tungsten. Molybdenum (Mo) and tungsten (W) are very similar chemically, although there are some differences that are not easy to explain. They have similar abundances and occur in nature in the same sorts of compounds; neither occurs as a free metal. For molybdenum the chief ore is molybdenite—molybdenum(IV) sulfide, MoS_2—but molybdates such as lead molybdate, $PbMoO_4$ (wulfenite), and $MgMoO_4$ are also found. Although tungsten occurs as tungstenite—tungsten(IV) sulfide, WS_2—the most important ores in this case are the tungstates such as scheelite (calcium tungstate, $CaWO_4$), stolzite (lead tungstate, $PbWO_4$), and wolframite—a solid solution or a mixture or both of the isomorphous substances iron(II) tungstate, $FeWO_4$, and maganese(II) tungstate, $MnWO_4$.

The small quantities of molybdenite in molybdenum ores are concentrated by foam flotation processes and roasted to give molybdenum(VI) oxide, MoO_3, which, after purification, can be reduced with hydrogen to the metal. For tungsten the ores are concentrated by magnetic and mechanical processes and the concentrate then fused with alkali. The crude melts are leached with water to give solutions of sodium tungstate, from which hydrous tungsten(VI) oxide is precipitated upon acidification, and the oxide is then dried and reduced to metal with hydrogen. The metals are quite refractory, the melting points being 2,610° C for molybdenum and 3,410° C for tungsten. Both metals are rather resistant to attack by acids, except for mixtures of concentrated nitric and hydrofluoric acids, and both can be attacked rapidly by alkaline oxidizing melts, such as fused mixtures of potassium nitrate and sodium hydroxide or sodium peroxide; aqueous alkalies, however, are without effect. The metals are inert to oxygen at normal temperature, but combine with it readily at red heat, to give the trioxides, and are attacked by fluorine at room temperature, to give the hexafluorides. Both elements form hard, refractory, and chemically inert interstitial compounds with boron, carbon, nitrogen, and silicon upon direct reaction with those elements at high temperatures. Tungsten carbide is extensively used for high-speed cutting tools; tungsten metal itself has wide application as a lamp filament, and molybdenum is used in a number of small-scale applications in which high resistance to erosion at high temperatures is required—*e.g.,* in jet engines, combustion liners, and afterburner parts.

Technetium and rhenium. As has been noted, technetium does not occur in nature; nor does rhenium occur free in nature or as a compound in any distinct mineral; instead it is widely distributed in small amounts in other minerals. Rhenium is generally isolated from flue dusts and gases obtained on the roasting of molybdenites that contain rhenium. Such gases can be leached or scrubbed with water to dissolve the oxide, Re_2O_7, which in turn can be converted to ammonium rhenate(VII) (perrhenate), NH_4ReO_4, and then reduced to the metal with hydrogen. The melting point of rhenium, 3,170° C, is exceeded only by those of tungsten and carbon.

Technetium and rhenium are very similar chemically and differ greatly from manganese. Their most characteristic and important compounds are formed in the oxidation states $+4$ and $+7$, although compounds are known in all formal oxidation states from -1 to $+7$. Rhenium has only minor uses; for example, as an alloy with tungsten in flash bulb filaments; technetium has essentially none.

Ruthenium and osmium. Both ruthenium and osmium have low crustal abundances of about 0.001 part per million. They occur in nature along with the other platinum metals, and processes for isolating them are an integral part of the metallurgical art that applies to all of them.

High resistance to chemical attack

Ruthenium and osmium are hard, bright metals having high melting points and resistance to chemical attack. They are the most noble of the platinum metals, and cold and hot acids, even including aqua regia, are without effect on them. They can be dissolved by fused alkalies, especially if an oxidizing agent such as sodium chlorate is present. Osmium will react at 200° C with air or oxygen to form the volatile tetroxide, OsO_4. Although ruthenium tetroxide, RuO_4, has similar stability and volatility to osmium tetroxide, OsO_4, it differs in that it cannot be formed from the elements.

The chemistries of ruthenium and osmium are generally similar and differ markedly from that of iron. The higher oxidation states $+6$ and $+8$ are much more readily obtained than for iron, and there is an extensive chemistry of the tetroxides, oxohalides, and oxo anions. For both elements there is little, if any, evidence that simple aquo ions exist, and virtually all their aqueous solutions, whatever the anions present, may be considered to contain complexes. In addition to carbonyl and organometallic compounds in the low oxidation states -2, 0, and $+1$, both elements form compounds in every oxidation state from $+2$ to $+8$. The most important oxidation states, however, are $+2$, $+3$, $+4$, $+6$, and $+8$.

Rhodium and iridium. Rhodium and iridium occur in nature in association with the other platinum metals, and their separation and refinement form part of the overall metallurgical processing of the group, which is discussed elsewhere. These metals are somewhat more reactive than ruthenium and osmium. Rhodium will dissolve in hot concentrated sulfuric acid and both dissolve in concentrated hydrochloric acid containing sodium perchlorate at 125°–150° C. Both will dissolve in aqua regia.

For both elements, the most important chemistry is that of the oxidation states $+1$ and $+3$, although iridium forms some complexes in oxidation state $+4$ and rhodium forms the dirhodium tetraacetate, $Rh_2(O_2CCH_3)_4$ and various derivatives containing two additional ligands—*e.g.,* water, pyridine, or triphenylphosphine—in oxidation state $+2$. Complexes in oxidation state $+1$ chiefly contain carbon monoxide, olefins, and phosphines as ligands. Hexachloroiridate(IV), $IrCl_6^{2-}$, and hexabromoiridate(IV), $IrBr_6^{2-}$, are the only notable chemical species containing tetravalent iridium.

Palladium and platinum. Platinum and palladium are more abundant than any of the other platinum metals. Platinum is by far the best known and technologically the most important. Both palladium and platinum are more reactive than the other platinum metals. Palladium is readily attacked by concentrated nitric acid and in its sponge form will dissolve even in hydrochloric acid in presence of chlorine or oxygen. Platinum is not attacked by any single mineral acid, but it dissolves readily in aqua regia and slowly in hydrochloric acid in presence of air. Both metals are rapidly attacked by fused alkali oxides and peroxides and also by fluorine and chlorine at about 500° C. They also combine with a number of nonmetallic elements on heating, such as phosphorus, arsenic, antimony, silicon, sulfur, and selenium.

Both metals, but especially palladium, are capable of absorbing large volumes of hydrogen. At 80° C and one atmosphere, palladium will absorb up to 900 times its own volume. The absorption causes both the electrical conductivity and magnetic susceptibility to decrease. These and other observations have led some investigators to conclude that there is a definite palladium hydride phase, although its structure and ideal composition are not firmly established.

Silver and gold. Silver and gold are two of the noblest—that is, least chemically reactive—of the transition elements, and both are found in nature in the free state. For silver the more important deposits commercially are such compounds as the mineral argentite (silver sulfide, Ag_2S), which is usually associated with other sulfides such as those of lead and copper, as well as several other sulfides, some of which contain antimony as well. Silver is found generally in lead ores, copper ores, and cobalt arsenide ores and is also frequently associated with gold in nature. Most silver is derived as a by-product from ores that are mined and processed to obtain these other metals. Gold is most commonly found free, the only compounds of gold in nature being various tellurides.

Silver is a white, lustrous metal and is among the most ductile and malleable ones known. Pure silver has the highest electrical and thermal conductivities of all metals. Gold is a soft, yellow metal that has the highest ductility and malleability of any element.

Reactivity of silver and gold

Silver is less reactive chemically than copper except toward sulfur and hydrogen sulfide, which rapidly blacken silver surfaces. The metal will dissolve in oxidizing acids and in solutions containing cyanide ions in the presence of oxygen or peroxides. Gold is much less reactive and is not attacked by oxygen or sulfur, although it will react readily with halogens or with solutions containing or generating chlorine, such as aqua regia. It also will dissolve in cyanide solutions in presence of air or hydrogen peroxide. Dissolution in cyanide solutions is attributable to the formation of the very stable dicyanoargentate(I), $Ag(CN)_2-$, and dicyanoaurate(I), $Au(CN)_2-$, ions.

As does copper, the elements silver and gold have a single *s* electron outside a completed *d* shell, but in spite of the similarity in electronic structures and ionization potentials there are few close resemblances between silver and gold on the one hand and copper on the other.

Both silver and gold find their chief applications in the form of the metals themselves, but silver is also of unique utility as silver chloride in photographic emulsions. Despite the relatively high cost of the material no satisfactory substitute has ever been found.

For silver the pre-eminently important oxidation state in all of its ordinary chemistry is the state +1, although the states +2 and +3 are known. With gold, on the other hand, the state +1 is generally quite unstable, and most of the chemistry of gold involves the state +3.

(F.A.C./Ed.)

Rare-earth elements

The rare-earth elements form a series of 17 chemically similar metals, all but one of which occur in nature. Often they are called simply rare earths, but this is a misnomer because the term earth properly is applied to the oxide of a metal rather than to the element itself. The rare-earth elements are not even particularly rare, though for a long time they were thought to be.

The 17 rare-earth elements are: scandium (Sc), yttrium (Y), lanthanum (La), cerium (Ce), praseodymium (Pr), neodymium (Nd), promethium (Pm), samarium (Sm), europium (Eu), gadolinium (Gd), terbium (Tb), dysprosium (Dy), holmium (Ho), erbium (Er), thulium (Tm), ytterbium (Yb), and lutetium (Lu).

Until the mid-20th century, there was not much use for pure rare-earth elements or compounds except cerium and lanthanum; mixtures of the rare earths, however, had found metallurgical and other uses. By the 1970s three of these elements, yttrium, gadolinium, and europium, were being used in the red phosphors for colour television.

Position in the periodic table

In the periodic table of the elements (see Figure 10), the rare-earth elements comprise three members of Group IIIb and all 14 members of one of two series of elements generally written apart from the main table. This long series is known collectively as the lanthanide series because it directly follows lanthanum in a different form of the table. The rare-earth elements all have certain common features in the electronic structure of their atoms, which is the fundamental reason for their chemical similarity.

The aqueous chemistry of all the rare earths is very similar and changes only slightly in progressing along the lanthanide series. Because of this similarity, it is difficult to separate individual rare earths. In the few cases in which the rare-earth ion can be oxidized or reduced to another valency, however, chemical separations can be carried out readily. Also, artificial mixtures of elements far apart in the series can be separated easily.

All of these elements form trivalent compounds, and in the crystal lattices (the regular arrangement of atoms in the solid forms) of such compounds, one rare-earth ion readily replaces another. The rare-earth metals when heated react strongly with nonmetallic elements to form very stable compounds. They are never found as the free metals in the Earth's crust. Pure minerals of individual rare earths do not exist in nature; all their minerals contain mixtures of the rare-earth elements.

Promethium is never found in the Earth's crust since it has no stable isotopes and is produced only by nuclear reactions; it can, however, be obtained in quantity from the fission products formed in nuclear reactors.

The chemical properties of scandium differ sufficiently from those of other rare-earth elements for it to have become segregated from them by the action of geological processes. Scandium seldom is associated with the rare earths in minerals.

HISTORY

The early Greeks defined earths as materials that could not be changed further by the sources of heat then available. Until late in the 18th century, this Greek conception remained strong in chemistry, and oxides of metals such as calcium, aluminum, and magnesium were known as earths and were thought to be elements.

Discovery of the first rare earth

In 1794, Johan Gadolin, a Finnish chemist, while investigating a rare Swedish mineral, discovered a new earth in impure form, which he believed to be a new element and to which he gave the name ytterbia, from Ytterby, the village where the ore was found. The name, however, was soon shortened to yttria. In 1803, from the same mineral, later named gadolinite in Gadolin's honour, another new earth was reported in the literature independently by several chemists. The new earth became known as ceria, from the asteroid Ceres, which had just been discovered (1801). Since yttria and ceria had been discovered in a rare mineral, and they closely resembled other known earths, they were referred to as the rare earths. Not until 1808 did Sir Humphry Davy demonstrate that the earths as a class were not elements themselves but were compounds of oxygen and metallic elements. Later, a number of chemists verified the existence of ceria and yttria in gadolinite and found that these oxides were also present in a wide variety of other rare minerals. The elements of which yttria and ceria were the oxides were then given the names yttrium and cerium, respectively.

In the period from 1839 to 1843, Carl Gustaf Mosander, a Swedish chemist (and student of Berzelius), found that yttria and ceria were not even the oxides of single elements but were, in fact, mixtures. He reported that if the oxides were dissolved in strong acid and the resulting solution subjected to a long series of fractional precipitations as various salts (oxalates, hydroxides, and nitrates), two new elemental substances could be split off from the main component of each oxide. The two new oxides found in ceria he called lanthana and didymia, and the elements contained in them were named lanthanum and didymium. The new elements found (as their oxides) in yttria he called erbium and terbium, and the oxides were referred to as erbia and terbia. Mosander also was the first to obtain the rare-earth metals themselves from their oxides, although only in impure form. Mosander's researches puzzled the scientists of his time. He seemed to be finding a new group of elements of an entirely different type from any known previously. All formed the same classes of compounds with almost the same properties, and the elements could be distinguished from one another—at that time—only by slight differences in the solubilities and molecular weights of the various compounds.

In the next few years the literature on the rare earths became confused. There was, for example, considerable controversy for a number of years over the existence of didymium. The situation was considerably clarified in 1859 when an instrument called the spectroscope was introduced into the study of the rare earths. This instrument indicated the patterns of light emission or absorption characteristic of the elements, and, with it, didymium was shown to have a characteristic absorption spectrum. From then on determination of spectra of various types became one of the most important tools in following the progress of the fractionation of rare earths. Somehow during this period the names used for the various fractions differed from laboratory to laboratory. Around 1860, by general agreement, it was decided to interchange the names of Mosander's earths, erbia and terbia.

Scandium and the periodic table

In 1869, when Mendeleyev first proposed the periodic table, he found it necessary to leave a blank at the position now occupied by scandium. He predicted, however, that a new element would be found to fit that blank in the table, and he also predicted certain properties of the element. The discovery of scandium a few years later (1879) and the agreement of its properties with those predicted by Mendeleyev helped to bring about general scientific acceptance of Mendeleyev's periodic table. Interestingly enough, one of the greatest weaknesses in the table was that it provided no logical place for the lanthanides, a difficulty that was not resolved for some years.

From 1843 to 1939 chemical fractionation of the mixed rare-earth salts obtained from many minerals was intensively investigated in both Europe and North America. Mosander's didymia was resolved into several oxides—samaria (samarium; 1879), praseodymia (praseodymium; 1885), neodymia (neodymium; 1885), and europia (europium; 1901). His terbia and erbia were resolved into holmia (holmium; 1878), thulia (thulium; 1879), dysprosia (dysprosium; 1886), ytterbia (ytterbium; 1876), and lutetia (lutetium; 1907).

During this period many of these elements were discovered independently by more than one investigator, but the credit for the discovery was usually given to the man who first separated sufficient quantities of the oxide to determine some of its properties and who published his results first.

As the scientists carried out their fractionations, they frequently observed changes in colour, apparent molecular weight, and spectra of the substances. Such changes were mainly responsible for the more than 70 claims for the discovery of new rare-earth elements during this period. Many of the observed changes were brought about by the concentration of different impurities, particularly the transition elements, in various fractions of the series. It is now known that such trace impurities in the rare-earth oxides can give rise to such colour changes and that such oxides can be made to fluoresce strongly and exhibit unique spectra.

Contributions of Welsbach

Shortly after Auer von Welsbach isolated praseodymia and neodymia in 1885 he invented an illuminating device that bears his name (Welsbach gas mantle), and a little later he produced a practical lighter flint. Both devices depended upon rare-earth elements. Although minerals rich in rare earths had up to that time been thought to be very rare, the demand for rare earths that developed as a result of Auer von Welsbach's inventions resulted in a worldwide search for rare-earth minerals, and it was found that one of them, monazite, existed in extensive deposits. Monazite, a phosphate of several rare-earth elements, was ideal for Auer von Welsbach's purposes, because it contained a high percentage of the element thorium, which was also used for the mantles. These were prepared by impregnating a cloth fabric with a solution of about 90 percent thorium nitrate, 10 percent cerium trinitrate, and minor amounts of other salts. When heated by a gas flame, these salts were converted to their oxides, which, when heated by the flame, gave off an intense white light. Cerium and iron form an alloy that emits sparks when struck. The discovery of this alloy by Auer von Welsbach started the flint industry. In 1913 about 3,300 tons of monazite were refined to produce the thorium and cerium used in gas mantles and the mixed rare-earth metals for flints and related products.

Atomic numbers of the rare-earth elements

The British physicist H.G.J. Moseley, while studying the X-ray emission spectra of the elements in 1913–14, found a direct relationship between the X-ray frequencies and the atomic numbers of the elements. This relationship made it possible to assign unambiguous atomic numbers to the elements and to verify their locations in the periodic table. In this way, Moseley was able to show clearly that there could be only 14 lanthanides, starting with cerium and ending at lutetium, and, at that time, all of the rare-earth elements had been discovered except for element 61. Because no stable isotopes (forms of the element with differing mass) of this substance exist in nature, it was not isolated until 1945, when one of its radioactive isotopes was separated from atomic fission products produced in a nuclear reactor. The element was named promethium after the Greek Titan who stole fire from the gods and gave it to mankind.

As in most fields of science, the present state of knowledge concerning the rare earths is the result of hundreds of scientists publishing thousands of papers and of the individual scientist making his advances based on the work that had been previously published. There were, of course, a number of men whose outstanding contributions changed the direction of the researches, but space does not permit referring to them by name.

OCCURRENCE AND ABUNDANCE

The rare-earth elements are not rare in nature. They are found in low concentrations widely distributed throughout the Earth's crust and in high concentrations in a considerable number of minerals. In addition, they are also found in many meteorites, on the Moon, and in the Sun. The spectra of many types of stars indicate that the rare-earth elements are much more abundant in these systems than they are in our solar system. Even promethium-147, which has a half-life (time required for one-half the material to undergo radioactive decay) of only a few years, has been observed in certain stars.

Cerium is reported to be more abundant in the Earth's crust than tin, and yttrium and neodymium more abundant than lead. Even the relatively scarce lutetium is said to be more abundant than mercury or iodine.

Widespread distribution on Earth

The rare-earth elements are found as mixtures in almost all massive rock formations, in concentrations of from ten to a few hundred parts per million by weight. The fact that these elements have not been separated into minerals containing individual members of the family at any time in the Earth's history—even after eons of repeated melting and resolidifying, mountain formation and erosion, exposure to hot vapour, and immersion in seawater—attests to the great similarity in properties of these elements. Nevertheless, rock formations resulting from some of these geological processes become enriched or depleted in rare earths at one end of the series or the other, so that an analysis of the relative content of the rare-earth elements is never exactly the same, even for similar rocks taken from different locations. In general, it has been found that the more basic (or alkaline) rocks contain smaller amounts of rare earths than do the more acid rocks, and it is believed that as these molten basic rocks intrude into the more acidic rocks, the rare earths are partially extracted into the more acidic rocks. Also, as this extraction takes place, the rare-earth elements of lower molecular weight (lanthanum, cerium, praseodymium, and neodymium) are taken up to a greater extent than the heavier elements.

Analytical methods involving activation analysis (production of artificial radioactivity) and mass spectroscopy (separation of atoms on the basis of mass) have made it possible to make accurate measurements of the relative abundances of these elements, even when they are present in extremely small amounts. Such measurements are of great interest to geophysicists because they supply valuable information about the development of geological formations. The cooling of molten rocks and superheated water solutions that have percolated through rock under great pressure frequently produces minerals containing up to 50 percent rare earths. (For uniformity, these percentages are calculated as if the entire rare-earth content of the mineral were present in the form of oxides.) From the presence and composition of such minerals, geochemists can learn a great deal about the conditions, such as temperature and pressure, to which the rock mass was subjected. Similarly, the relative abundance of rare earths in the rocks on the Moon is of great interest because of what it is expected to reveal about how the Moon was formed and whether all or part of the Moon was molten at any time.

The average content of rare-earth elements found in certain meteorites (chondrites) and in three types of common rocks is listed in Table 25. Included also is an estimate of the relative abundance of the elements in terms of the overall rank of all known elements and of their concentration in the Earth's crust. It is now generally accepted that the relative values of the rare-earth elements in chondritic (granular) meteorites represent their overall relative abundance in the Cosmos. The elements with even atomic numbers are much more abundant than the odd-numbered elements. Such information, together with the relative abundance of their isotopes, is of critical importance to astrophysicists because it bears on theories of the origin of the universe and the genesis of the chemical elements.

GENERAL PROPERTIES OF THE GROUP

Location of lanthanide elements in the periodic table

Electronic structure. The 14 lanthanide elements—from cerium through lutetium—are much alike because the differences in their electronic structures chiefly involve the inner 4*f* electrons, whereas it is the outer *s* and *p* (and sometimes *d*) electrons that are involved in chemical bonding with other atoms and thereby determine the chemical behaviour of the elements. Although lanthanum atoms contain no 4*f* electrons, they resemble the atoms of the lanthanide elements closely, and it is not surprising that lanthanum should behave much as the lanthanides do (the name lanthanide, in fact, merely means lanthanum-like). Scandium and yttrium are elements in the same vertical file in the periodic table as lanthanum, and their atoms, too, have somewhat the same electronic structure but fewer filled shells, the outermost electrons in scandium being two 4*s* electrons and one 3*d*. In the case of yttrium, however, the outermost electrons are 5*s* and 4*d* electrons, respectively.

Because of their general similarity in atomic structure, scandium, yttrium, lanthanum, and the 14 lanthanides are very similar chemically. This similarity is the reason they are found together in nature and also the reason they are so frequently classed together as the rare-earth elements.

Table 26 lists the lowest energy electronic configurations of the rare-earth elements in gaseous and in condensed states, *i.e.*, as the free metal or in a crystalline or solution form as a compound. In a compound the $5d6s^2$ electrons (the superscript indicating the number of electrons in the subshell) are the valence, or bonding, electrons. These electrons are involved in the chemical bond and are usually paired with the electrons of the anion (the negative ion included with the rare-earth ion in the compound), with the result that they are no longer closely associated with the rare-earth atom. In the case of the metal, these electrons are free to wander throughout the crystal, being able to carry an electrical current and known, therefore, as conducting electrons. In the case of lanthanides in the gaseous state, the valence electrons are localized at the individual atoms, but in many cases the 5*d* electron can switch to the 4*f* subshell, giving rise to a state of different energy and different electronic configuration.

Lanthanide contraction

As the charge on the nucleus increases across the lanthanide series, it pulls the various subshells, especially the 5*s* and 5*p* subshells, closer to the nucleus. As a result, the radii of the lanthanide ions decrease as the atomic number increases. This effect is known as the lanthanide contraction.

Physical properties. The pure rare-earth metals are bright and silvery. A bar of europium will tarnish almost immediately when exposed to air and will be entirely converted to the powdered oxide in a few days. Lanthanum, cerium, praseodymium, and neodymium also corrode readily in air; bars of these metals become encrusted with a thick layer of oxide in several weeks. Metallic yttrium, gadolinium, and lutetium, on the other hand, remain bright and shiny for years.

The properties of the rare-earth metals are frequently quite sensitive to the presence of impurities; for example, the light lanthanide metals will corrode much more rapidly if small amounts of calcium or magnesium or rare-earth oxides are present in the metal. The melting points and transition temperatures between different crystal forms (allotropic forms) can be changed drastically, frequently by several hundred degrees, when the metals are alloyed with other elements.

Table 25: Abundance of the Naturally Occurring Rare-Earth Elements
(parts per million)

element	earth's crust		average of 20 chondritic meteorites	composite of 40 North American shales	western North American Precambrian granites	Kilauea basalt
	rank*	abundance				
Sc	46	5.0	—	—	—	—
Y	31	28.0	1.800	35.00	31.00	—
La	35	18.0	0.300	39.00	49.00	10.50
Ce	29	46.0	0.840	76.00	97.00	35.00
Pr	45	5.5	0.120	10.30	11.00	3.90
Nd	32	24.0	0.580	37.00	42.00	17.80
Sm	42	6.5	0.210	7.00	7.20	4.20
Eu	57	1.1	0.074	2.00	1.25	1.31
Gd	43	6.4	0.320	6.10	5.80	4.70
Tb	59	0.9	0.049	1.30	0.94	0.66
Dy	50	4.5	0.310	—	—	3.00
Ho	56	1.2	0.073	1.40	1.22	0.64
Er	54	2.5	0.210	4.00	3.20	1.69
Tm	65	0.2	0.023	0.58	0.53	0.21
Yb	53	2.7	0.170	3.40	3.50	1.11
Lu	60	0.8	0.031	0.60	0.52	0.20

*Expressed in a range of 1 to 105.

Table 26: Some Properties of the Rare-Earth Elements (values recommended by Ames Laboratory)

	Scandium	Yttrium	Lanthanum	Cerium	Praseodymium	Neodymium	Promethium	Samarium
Atomic number	21	39	57	58	59	60	61	62
Atomic weight	44.956	88.905	138.91	140.12	140.907	144.24	(145)*	150.35
Colour of element	silvery	silvery	silvery	silvery	silvery	silvery	silvery	silvery
Melting point (°C)†	1,541	1,522	921	799	931	1,021	1,168	1,077
Boiling point (°C)¶	2,831	3,338	3,457	3,426	3,512	3,068	2,700	1,791
Density at 25° C (g/cm³)	2.9890	4.4689	6.1453	6.672	6.773	7.007	—	7.520
Conduction electrons in metal	3	3	3	3 (3.1)	3	3	3	3
Valence in aqueous solution	3	3	3	3,4	3	3	3	3,2
Colour in aqueous solution	colourless	colourless	colourless	colourless	green	rose	colourless	yellow
Colour of oxide	white	white	white	off-white (Ce_2O_3); cream (CeO_2)	black (Pr_6O_{11}); green (Pr_2O_3)	blue		cream
Electronic configuration	[Ar] $3d^14s^2$	[Kr] $4d^15s^2$	[Xe] $5d^16s^2$	[Xe] $4f^26s^2$	[Xe] $4f^36s^2$	[Xe] $4f^46s^2$	[Xe] $4f^56s^2$	[Xe] $4f^66s^2$
Heat of fusion (kcal/g-atm)	3,369	2,732	1,482	1,238	1,652	1,705	—	2,061
Heat of vaporization¶ (kcal/g-atm)	89.9	101.3	103.1	101.1	85.3	78.5	—	49.2
Electrical resistivity at 25° (ohm-cm × 10^{-6})	52	59	61–80	70–80	68	65	—	91
Nèel point (°K)	none	none	none	12.5	—	20	—	15
Curie point (°K)	none	none	none	—	—	—	—	—
Compressibility at 25° C (cm²/kg × 10^{-6})	2.26	2.68	4.04	4.10	3.21	3.00	(2.8)	3.34
Crystal structure‡	hcp	hcp	d-hcp	fcc	d-hcp	d-hcp	—	rhomb
Radius, metallic (Å)	1.640	1.801	1.879	1.820	1.828	1.821	—	1.804
Radius, ionic (Å)	0.732 (M^{3+})	0.893 (M^{3+})	1.061 (M^{3+})	1.034 (M^{4+}, 0.92)	1.013 (M^{4+}, 0.90)	0.995 (M^{3+})	0.979 (M^{3+})	0.964 (M^{2+}, 1.11)

	Europium	Gadolinium	Terbium	Dysprosium	Holmium	Erbium	Thulium	Ytterbium	Lutetium
Atomic number	63	64	65	66	67	68	69	70	71
Atomic weight	151.96	157.25	158.9254	162.50	164.930	167.26	168.934	173.04	174.97
Colour of element	silvery	silvery	silvery	silvery	silvery	silvery	silvery	silvery	silvery
Melting point (°C)†	822	1,313	1,356	1,412	1,474	1,529	1,545	819	1,663
Boiling point (°C)¶	1,597	3,266	3,123	2,562	2,695	2,863	1,947	1,194	3,395
Density at 25° C (g/cm³)	5.2434	7.9004	8.2294	8.5500	8.7947	9.066	9.3208	6.9654	9.8404
Conduction electrons in metal	2	3	3	3	3	3	3	2	3
Valence in aqueous solution	3,2	3	3	3	3	3	3,2?	3,2	3
Colour in aqueous solution	colourless	colourless	colourless	yellow tint	yellow	pink	greenish tint	colourless	colourless
Colour of oxide	white, greenish tint	white	brown (Tb_4O_7); white (Tb_2O_3)	white	yellowish white	pink	white, greenish tint	white	white
Electronic configuration	[Xe] $4f^76s^2$	[Xe] $4f^75d^16s^2$	[Xe] $4f^96s^2$	[Xe] $4f^{10}6s^2$	[Xe] $4f^{11}6s^2$	[Xe] $4f^{12}6s^2$	[Xe] $4f^{13}6s^2$	[Xe] $4f^{14}6s^2$	[Xe] $4f^{14}5d^16s^2$
Heat of fusion (cal/g-atm)	2,204	2,403	2,583	2,577	4,033	4,757	4,025	1,830	(4,457)
Heat of vaporization¶ (kcal/g-atm)	(41.9)	95.3	93.4	70.0	72.3	76.1	55.8	36.5	102.2
Electrical resistivity at 25° (ohm-cm × 10^{-6})	91	127	114	100	88	71	74	28	60
Nèel point (°K)	90	—	229	179	132	85	57	none	none
Curie point (°K)	—	293.2	221	87	20	20	. . .	none	none
Compressibility at 25° C (cm²/kg × 10^{-6})	8.29	2.56	2.45	2.55	2.47	2.39	2.47	7.39	2.38
Crystal structure‡	bcc	hcp	hcp	hcp	hcp	hcp	hcp	fcc§	hcp
Radius, metallic (Å)	1.984	1.801	1.783	1.774	1.766	1.757	1.746	1.939	1.735
Radius, ionic (Å)	0.950 (M^{2+}, 1.09)	0.938 (M^{3+})	0.923 (M^{4+}, 0.84)	0.908 (M^{3+})	0.894 (M^{3+})	0.881 (M^{3+})	0.870 (M^{3+})	0.858 (M^{2+}, 0.93)	0.850 (M^{3+})

*Radioactive isotope separated from fission residues. †Corrected for new temperature scale. ‡hcp = hexagonal close packed; fcc = face-centred cubic; rhomb = rhombohedral; bcc = body-centred cubic. §Structure normally found at 25° C; other forms known. ¶R. Hullgren, R. Orr, and K. Kelley.

Small amounts of nonmetallic impurities also effect many of the properties of the rare-earth elements. Several thousand parts per million by weight of oxygen and even smaller amounts of nitrogen in the metals make them brittle. The effect of nonmetallic impurities on physical properties is determined by atomic percentages (that is, by the relative numbers of atoms present) and not by weight percentages; thus, in lutetium 300 parts per million by weight of hydrogen is about 5 percent on an atomic basis, whereas 1,000 parts per million of oxygen in lutetium by weight represents only 1 percent oxygen on the atomic scale.

In determining properties of the rare-earth metals it is obviously essential to work with well-characterized samples. The amount of each individual impurity present should be accurately known, as well as the previous history of the sample with regard to temperature and work. If the reported values of physical properties are to have much meaning, this information concerning the samples used must be given. Unfortunately, because of a lack of appreciation of the importance of impurities or a lack of proper equipment to adequately characterize the sample, there is a wide variation in the numbers reported in the literature for certain properties. In Table 26 are the values of the properties that—in the author's opinion—are the best values known. Some properties, such as elastic constants, resistivity, and effective magnetic moments, are very sensitive to temperature and show marked anomalies at and in the neighbourhood of crystal or magnetic transformations. Also, some properties depend on the angle at which they are measured with respect to the principal crystal axes in the metal.

In fact, the rare-earth metals do not resemble one another as closely as was generally believed in the early part of the 20th century. Physical properties differ as much across the lanthanide series as they do for most other series in the periodic table. The melting point of lutetium, for example, is almost twice that of lanthanum, and the vapour pressures of ytterbium and europium at 1,000° C are millions of times greater than those of lanthanum and cerium. Lanthanum is a superconductor of electricity at 6 K (−267° C) and gadolinium is a stronger ferromagnet at 0 K than iron. The properties of adjacent pairs of lanthanide elements do, however, differ in a predictable and usually regular manner. This behaviour makes these metals uniquely valuable in studying theories of the metallic state, the formation of alloys, and the existence of intermetallic compounds. Each of the rare-earth metals readily combines with almost any other metallic element, and the resulting alloys exhibit a wide variety of properties: they can be hard or soft, brittle or ductile, and they can have high or low melting points. Some are extremely pyrophoric (ignite spontaneously), whereas others cause a coating to be formed on the surface of metals such as magnesium that protects the alloys from corrosion at elevated temperatures.

Rare-earth metals absorb hydrogen to form stable alloylike hydrides in which percentages of the divalent compounds MH_2 range from zero to 100. These hydrides, brittle and metallic in appearance, have a bluish tinge. After absorption of hydrogen to yield the composition MH_2, further absorption occurs, finally yielding trivalent hydrides MH_3. During the change from MH_2 to MH_3, the properties become more saltlike. The amount of hydrogen per unit volume in yttrium hydride is considerably greater than that in water or liquid hydrogen, and this hydride does not develop a partial pressure of hydrogen gas equal to one atmosphere until the alloy has been heated to a white heat. Cerium metal, once the oxide surface film has been broken, absorbs hydrogen at room temperature and decomposes water vapour at higher temperatures, absorbing the hydrogen and reacting with the oxygen to form a layer of sesquioxide on the surface. The oxides, nitrides, and carbides of the rare-earth elements are soluble in the molten metals, as are the elements oxygen, nitrogen, and carbon. The exact form in which the dissolved substances are present is not known, but it is generally believed that the nonmetallic elements are present as interstitial atoms (atoms inserted in spaces left in the crystal structure). These dissolved nonmetallic elements remain in solid solution over a considerable composition range at temperatures near the melting point. As the metal is slowly cooled, however, the solubility decreases, and the dissolved elements precipitate as a second phase, probably as the sesquioxides, nonstoichiometric nitrides, and carbides. The diffusion rate (rate of movement) for nonmetallic elements in the metal is low below 800° C and becomes progressively lower as the temperature is lowered. The properties of the metals containing these impurities, therefore, are dependent upon the heat treatment to which they have been subjected.

Heat
reatment
and crystal
structure

Anisotropy. Rare-earth metals often exhibit anisotropy—differences in properties depending on which direction in the crystal they are measured—and the heat treatment of the sample is important in producing polycrystalline metal. It is possible by certain heat treatments to produce large grains oriented preferentially in a given direction. When properties that depend on crystal direction are measured on such polycrystalline samples, the results have little meaning unless the amount of preferred orientation is known.

Energy states of lanthanides. The spark and arc spectra (patterns of emitted light) of the gaseous lanthanides are extremely complicated. There are literally tens of thousands of frequencies of light emitted by each of the lanthanides, and it requires very powerful instruments (spectrographs) to resolve them. This complexity arises from the fact that the lanthanides have an incomplete inner subshell, and the angular moments (spins and orbital motions) of the electrons in this subshell can combine in many ways to give many different energy states. In the most complicated case, that of gadolinium, there are 3,432 different states. Any of these states can combine with the many states arising from the three valence electrons, and this condition results in an incredible number of excited energy levels. The emitted frequencies represent transitions between any of these states. The situation is further complicated by the fact that the ionization potentials of these elements are extremely low, with the result that in the arc- and spark-light sources there are a great many ionized atoms, and their complicated spectra fall on top of those of the neutral atoms. Many thousands of these lines remain to be identified. A start has been made on this task, however, and a few levels—including the basic ones—have been identified. A thorough understanding of the energy levels of the rare-earth atoms will be of great value in arriving at a complete theory by which all the properties of an atom can be calculated from basic principles. Furthermore, the complete identification of the spectral lines of the rare-earth elements will be of great assistance to astronomers in identifying the many lines observed in stellar spectra that are believed to indicate the presence of rare-earth elements.

The first ionization potentials (the energy required to remove an electron from the neutral atom) of these elements are also difficult to determine accurately because of the complexity of the rare-earth spectra.

Rare-earth ions in solids and liquids. The sharp bands in spectra of solid rare-earth elements and compounds are much better understood. These bands arise from transitions between different energy states of the $4f$ subshell, and the position of the bands seems to be little affected by the atoms surrounding the lanthanide atoms. For this reason, scientists have been able to use these bands for more than a century to determine whether particular rare-earth ions are present in solids or liquids. The fine line structure of the bands, which can be resolved at low temperatures, is sensitive to the environment, and this effect makes such spectra a valuable tool for studying the forces that exist in solids and liquids.

Magnetic
properties

The $4f$ electrons also are responsible for the strong magnetism exhibited by the metals and compounds of the lanthanides. In the incomplete $4f$ subshell the magnetic effects of the different electrons do not cancel out each other as they do in a completed subshell, and this factor gives rise to the interesting magnetic behaviour of these elements. At higher temperatures, all the lanthanides except lutetium are paramagnetic (weakly magnetic), and this paramagnetism frequently shows a strong anisotropy. As the temperature is lowered, many of the metals exhibit a point below which they become antiferromagnetic (*i.e.,* magnetic moments of the ions are aligned but some are opposed to others), and, as the temperatures are lowered still further, many of them go through a series of spin rearrangements, which may or may not be in conformity with the regular crystal lattice. Finally, at still lower temperatures, a number of these elements become ferromagnetic (*i.e.,* strongly magnetic, like iron). Some of the metals have saturation moments (magnetism observed when all the magnetic moments of the ions are aligned) greater than iron, cobalt, or nickel. They also show a strong anisotropy in their magnetic behaviour depending on the crystal direction. Study of the magnetism of rare-earth elements has had great influence on present-day theories of magnetism.

Electrical conductivity. The rare-earth metals, with the exceptions of cerium, ytterbium, and europium, have three electrons available for carrying electrical current. The space occupied by these electrons apparently represents more than 85 percent of the volume associated with the atom of each metal. Cerium is reported to have an average of 3.1 conducting electrons, presumably as the result of the existence of some of its atoms in a state in which four electrons are free to move through the metal. Pure cerium under high pressure or at low temperature assumes a high-density form in which the four-electron state assumes more importance. Europium and ytterbium are much less dense than the other lanthanides, and they have only two conducting electrons; the third valence electron has moved to an inside subshell ($4f$). In europium this electron half fills this subshell, and in ytterbium it completes it, the two configurations $4f^7$ and $4f^{14}$ being particularly stable. The electrical and chemical properties of these two metals therefore resemble those of magnesium, calcium, strontium, and barium (metals with two conducting electrons) more closely than those of the other lanthanides.

Nuclear properties. As a group the rare-earth elements are rich in total numbers of isotopes, averaging about 20 each. The elements with odd atomic numbers have only one, or at most two, stable isotopes, but those with even atomic numbers have from four to seven stable isotopes. Some of the unstable isotopes are feebly radioactive, having extremely long half-lives. The unstable radioactive isotopes can be produced in many ways; *e.g.,* by fission, neutron bombardment, radioactive decay of neighbouring elements, and bombardment of neighbouring elements with charged particles. The lanthanide isotopes are of particular interest to nuclear scientists because they offer a rich field for testing theories about the nucleus, especially because many of these nuclei are nonspherical, a property that has a decided influence on nuclear stability. When either the protons or neutrons complete a nuclear shell (that is, arrive at certain fixed values), the nucleus is exceptionally stable—the number of protons or neutrons required to complete a shell being called a magic number. One particular magic number—82 for neutrons—occurs in the lanthanide series.

PRODUCTION AND APPLICATION

Monazite

Sources and extraction. Though numerous minerals rich in rare earths are found in the Earth's crust, many are extremely rare, and many more are found only in small pockets in more massive rocks. Although such minerals are of considerable research interest they are not used commercially. Monazite, a mixed phosphate of calcium, thorium, cerium, and various lanthanides, occurs in extensive deposits and is one of the main sources used commercially to obtain the light rare-earth elements. Monazite contains about 50 percent by weight rare-earth elements, in the approximate proportions 50 percent cerium, 20 percent lanthanum, 20 percent neodymium, 5 percent praseodymium, and lesser amounts of samarium, gadolinium, and yttrium. It also contains small amounts of the heavy rare-earth elements. The actual amounts of each element in the mineral vary considerably, depending on the point of origin of the monazite, because the various metallic elements can substitute for one another in the crystal lattice. The mineral probably formed as small crystals in rocks as they cooled, but as the mountains eroded away and were washed into the sea, the monazite, being denser than most other materials, settled first, while the lighter materials were carried farther out to sea. Apparently as a result of this action, sandbars containing monazite are found along the coasts of Brazil and southwest India. Concentrated deposits are also found on certain uplands, which are thought to have been the beaches of ancient seas or oceans and which were later uplifted. Such deposits in massive amounts are found in Australia, the U.S.S.R., and South Africa, and in the United States in South Carolina, Florida, and Idaho, and in many other locations. The mineral is dredged or scooped up, pulverized if necessary, and concentrated by flotation methods. Sometimes a magnetic-belt separator is used to pull the more magnetic monazite to one side in order to separate it from the nonmagnetic materials. The monazite is then shipped to rare-earth chemical plants. The mineral xenotime, a phosphate of yttrium and various lanthanides, is frequently found associated with monazite and may comprise from 1 to 10 percent of the mixed minerals. It is similar to monazite except that the metallic atoms are about 50 to 60 percent yttrium, and it contains more heavy lanthanides than light ones. Xenotime is one of the main sources of the heavy rare earths, and it can be separated from monazite by the magnetic-belt process because it is more magnetic than monazite.

Another important source of light rare earths and europium is the mineral bastnaesite, a fluorocarbonate of lanthanum and cerium, with smaller amounts of neodymium and praseodymium. It is found in extensive deposits in eastern California. It contains almost no heavy rare earths, but there is enough europium (about 0.1 percent) to supply much of the world demand for this element. The mineral is also found in the U.S.S.R. and in parts of Africa. The rock is broken up by blasting and then is crushed and ground to a fine powder. The bastnaesite is separated from the other materials by the usual flotation methods and is then treated chemically so that it can be separated into europium, lanthanum, and cerium fractions by liquid–liquid extraction methods (see below *Liquid–liquid extraction*).

The niobium titanate minerals, such as fergusonite, euxenite, samarskite, and blomstrandine, are rich in the heavy rare-earth elements but are not used much commercially. The same is true of such silicates as gadolinite and allanite. Other commercial sources of rare-earth oxides are certain uranium- and apatite-mining operations in which the rare earths are obtained as a by-product even though the rare-earth content of the ores is low.

Very little scandium is found in rare-earth minerals. Most of the scandium produced commercially is a by-product from uranium processing—the scandium, which may be present in amounts up to five parts per million, being recovered from the uranium solution. There is, however, a rare mineral thortveitite—found in Norway—that contains up to 34 percent scandia, Sc_2O_3.

Purification of europium

Methods of separation and purification. Generally, the rare-earth elements exist in dilute solution as trivalent ions. Quite early, however, it was found that a number of the elements could also exist in tetravalent or divalent form—including cerium(IV), samarium(II), europium(II), and ytterbium(II). If an element could be oxidized (to the tetravalent state) or reduced (to the divalent), then it could be removed readily from the other rare earths. Between the years 1930 and 1935, for example, about two kilograms of extremely pure europium compounds were prepared by a separation process making use of the divalent form of europium. Although europium is one of the less abundant rare-earth elements, it was one of the first of the heavier rare earths to become generally available.

Fractionation. Because the ions of the rare-earth elements are surrounded by tightly bound water molecules in aqueous solution, compounds of the rare earths formed from aqueous solutions have properties much alike, and this similarity is particularly true for adjacent elements. The problem is still further complicated by the fact that one rare-earth ion can be substituted readily for another in crystal lattices, with the result that most precipitates consist of crystals of almost the same rare-earth mixture as the solution. Because of this behaviour, chemists of the 19th and early 20th centuries found it necessary to resort to laborious fractionation processes to isolate individual rare-earth elements. At the time, many different processes were used, such as fractional crystallization, fractional precipitation, fractional decomposition, and fractional extraction. All of these consisted of separating the mixed rare earths into two approximately equal fractions, one of which would be enriched in the lighter elements and the other in the heavier elements. Both fractions would then be put back into solution and the process repeated on each of them. Usually the adjacent inner fractions would be recombined before proceeding to the next stage. Gradually, the lighter rare earths were collected in the beakers toward one end of the system, with the heavier elements concentrated at the other end.

As the quantity of material in the end beakers became small, it was usually customary to combine equivalent fractions from other similar runs. At this point the first series would be split into several independent groups, and a new fractionation process more suited to the elements in each fraction started. Needless to say, the quantity of a relatively pure rare-earth compound obtained from the end beakers was distinctly limited.

Fractional separation methods

Fractional separation methods, particularly for adjacent heavy rare earths, are extremely slow and tedious. One investigator, for example, reported that he had recrystallized the bromate salt of a thulium fraction 15,000 times and could see little difference between the first and last fractions. It is now known that even the purest fractions he obtained contained some ytterbium and erbium. If the elements are far apart in the lanthanide series, however, the task is simplified. It takes only a few partial precipitations, for example, to obtain a lanthanum–cerium–praseodymium fraction completely free of erbium, thulium, ytterbium, and lutetium. The most basic (nonacidic) of the rare-earth elements, lanthanum, is very well situated in the series in this respect because it occurs at one end, and a few fractionations suffice to separate a lanthanum–cerium fraction from the other rare earths.

Since cerium in its tetravalent form has distinctly different chemical properties from a typical lanthanide in the trivalent state, it can be separated from lanthanum easily by ordinary chemical operations. Consequently, pure lanthanum and cerium compounds have been commercially available for many years, and even today several companies find the fractionation process the most economical method for producing compounds of these elements in ton quantities.

Ion exchange of fission products

Ion exchange. Ion exchange is a method of separation based on differential absorption and elution (washing off) of substances from certain solid supporting materials, often powdered or finely divided materials held in glass tubes. The technique was first used in the rare-earth field during World War II to separate fission products obtained from nuclear reactors. In December 1943 a research group at the Oak Ridge (Tennessee) national laboratory announced that they had separated the mixed rare-earth elements from certain fission products by ion exchange on an organic resin into three fractions. The first fraction was shown to have radioactivity associated with yttrium, and the final peak to have cerium activity. The middle peak was thought to be a combination of the neodymium and element-61 activities. The group at Oak Ridge continued to develop the elution technique for separating fission products both with and without carriers (nonradioactive materials added to carry with them the radioactive isotopes). By the end of the war, they had succeeded in developing the processes so that they could separate the individual rare-earth elements of the cerium group (cerium through element 61) and yttrium. The carriers usually consisted of a few milligrams of each of the corresponding natural rare earths. In the meantime, a group at Iowa State University applied the ion-exchange process to the separation of gram quantities of adjacent rare earths and succeeded in separating the difficult pair praseodymium and neodymium in fairly high purities in gram quantities.

The ion-exchange elution method

Rare-earth-element separation by the ion-exchange elution process is carried out as follows: At the start of the process, the resin is saturated with monovalent cations, such as ammonium ion, NH_4^+, or hydrogen ion, H^+. Next, a solution of mixed rare-earth ions accompanied by strong acid anions is poured onto the top of the column. When the rare-earth ion encounters the cation-containing resin, it replaces three monovalent cations, and these—along with the strong acid anions—will flow through the column in solution and out the bottom. A band of resin saturated with rare-earth ions forms at the top of the ion exchanger and grows in length as more rare-earth solution is added. There is, however, little separation of individual rare-earth ions as this band forms. An eluant solution containing an anion that complexes with the rare-earth ion is then prepared, for example, an ammonium citrate solution of controlled acidity. This solution is then started flowing through the column to elute the rare-earth band down the column and out the bottom. When the main ions present in ammonium citrate in acid solution, $HCit^{2-}$ or H_2Cit^-, encounter rare-earth ions on the resin, complex ions form; these enter the solution phase, and three monovalent ions deposit on the resin in their place. When the rare-earth complexes reach the ammonium or acid resin, in front of the rare-earth band, the rare-earth ions are again deposited, and the band progresses down the column. The formation constants of the individual rare-earth complexes increase slightly with increasing atomic number. Because the various rare-earth ions on the resin are in equilibrium with the rare-earth complexes in solution as they pass over the band, there is a slight enrichment of heavy rare earths at the front of the band. As the band progresses down the column, this enrichment continues. At the same time, the band grows in length, since ammonium and hydrogen ions are also in equilibrium with the resin and their ions will deposit along with the rare-earth ions. After the band has travelled many band lengths, each rare earth exhibits a bell-shaped elution curve (concentration of rare-earth ions versus volume of eluant leaving the column) and these individual rare-earth bands travel down the column at different rates. The bands overlap badly at first, but after travelling many band lengths, they pull completely apart. The area under each curve is, of course, constant, because the amount of each rare earth on the column does not change, but the concentration of the rare earth in the resin gets less and less relative to the ammonium and hydrogen ions on the resin.

Limitations of the elution technique

With this type of separation, the original mixed rare-earth band must be quite narrow because the band has to travel many band lengths on a given column. The ions in the eluant are constantly overrunning the bands, with the result that large quantities of solution are needed; and the solution coming out the bottom of the column containing the successive pure rare earths is extremely dilute in rare earths. Such a process is obviously ideal for separating radioactive tracers, which one can count by means of radioactivity, and this process is frequently used in analytical chemistry, where only small amounts of the rare earths are separated. When it is necessary to obtain large amounts of rare earths in high purity, this process is not effective. It has the disadvantage of requiring far more chemicals than the displacement method developed later and described below. Furthermore, this process is not particularly adaptable to being scaled up to produce large quantities of ultrapure rare earths, nor is it well suited for recycling the water and chemicals. It does not give the purity of the individual rare earth that displacement methods can achieve. Finally, the elution process is slow compared with the displacement method.

The ion-exchange displacement method

The band displacement method of separating individual rare-earth elements was first published in 1952. This process is capable of being scaled up to handle any quantity of rare earths. The mixture can be resolved so that 98 or 99 percent of each individual rare earth can be recovered with less than 0.1 percent of other rare-earth impurities; and, if the rare earths are taken from the middle third of the bands, the sum total of other rare earths can be kept as low as 0.0001 percent. The same resins and type of equipment are used in this process as in the elution technique. Two strong chemical constraints, however, are imposed at the top and bottom edges of the rare-earth band. The eluant contains a strong complexing ion—such as a chelating agent, an organic molecule that wraps itself around the rare-earth ion, replacing all or most of the adjacent water molecules. The first constraint requires that the formation constants of the rare-earth complexes formed should be large enough so that, when the chelating agent encounters the top edge of the rare-earth band, it complexes in a short distance all of the rare-earth ions, moving them into solution and replacing them with the cation of the eluant. (The formation constant, however, should not be so large as to remove all the rare-earth ions from the solution phase.) The second chemical constraint occurs at the bottom edge of the rare-earth band: the original resin bed, called the retaining bed, down which the rare-earth band is moving, must have cations on its exchange sites that form a much tighter soluble complex with the chelating ion than do the rare-earth ions. Under this constraint the rare-earth complex promptly breaks up at the point where it encounters the retaining bed, and the rare-earth ions completely deposit in the bed, simultaneously removing an equivalent amount of the retaining-bed cations. With these constraints, the rare-earth band, after spreading out slightly to reach equilibrium, remains of constant length, with sharp top and bottom edges, no matter how far down the column it travels. The elution curve is flat-topped (rare-earth concentration remains constant over almost the entire band when plotted against volume of elute leaving the bottom of the column), and the percentage of rare-earth ions in the rare-earth band on the active sites of the resin is close to 100 percent. Here again, there must be a slight difference in the formation constants of the rare-earth chelates, so that the rare-earth ions are constantly interchanging as the eluant flows by the rare-earth band. As the band moves, the individual rare earths separate into individual flat-topped bands, which ride head to tail and never pull apart. By the time the band has travelled a tenth of its length, most of the heavy rare earths are already to be found in the front segments of the total rare-earth band, and, by the time it has travelled one length, all the individual rare earths are in separate bands, which overlap only slightly to give a narrow region consisting of a binary mixture of adjacent elements. These mixed regions, of course, must be recycled. By having a series of columns, however, the original rare-earth band can be made very long, and, since the overlap regions are independent of band length, the bulk of each successive rare earth comes out the bottom of the column in high purity.

A number of companies have adopted the displacement process and, using it, have made available highly pure salts of the rare-earth elements of atomic number 59 and above (all the elements from praseodymium through lutetium) in any quantity at reasonable prices. This process has the distinct advantage of allowing the water and the eluting chemicals to be recycled and used over again. One long absorbed band can follow another down a series of columns if a short retaining-bed section is continually regenerated between the absorbed bands.

The first chelating agent used was ammonium citrate at such a low acidity that the citrate ion, Cit^{3-}, predominates in the solution. At this acidity the complex chelate ion, $MCit_2^{3-}$, forms. This process works well, but in 1954 it was improved by using a buffered ammonium solution of ethylenediaminetetraacetic acid (EDTA) with a cupric-ion retaining bed. A number of other chelating agents and types of retaining beds have also been investigated. Many of these work well, but none is markedly superior to EDTA. Some of these other systems, however, are used for certain regions of the rare-earth series because, for these regions, the processing is accomplished more quickly and cheaply.

Organic and aqueous phases

Liquid–liquid extraction. Liquid–liquid extraction methods also find important applications in the rare-earth industry. The basic principles involved are similar to those operating in the ion-exchange processes. An organic solvent, such as tributyl phosphate, flows countercurrent to an aqueous stream containing the mixed rare-earth salts. Rare-earth complexes are formed with formation constants that vary somewhat across the series. The rare-earth ions can complex with their own anions to form neutral molecules that are soluble in the organic phase, or they can complex with molecules of the organic solvent and thereby join the organic stream. If desired, an organic chelating agent can be added to form complexes with the rare-earth ions. These complexes should be soluble in the organic liquid. As the aqueous phase flows past the immiscible organic stream, an equilibrium is set up between the rare-earth ions in the aqueous solution and the complex ions in the organic solvent. As the two streams flow past each other, the heavy rare-earth elements concentrate in one stream and the lighter ones in the other.

The equilibrium constant for the exchange of one rare-earth ion for another is usually small, with the result that the ions have to exchange with the complex many times before a clear-cut separation between two rare-earth ions is achieved. This process necessitates that the two liquids be in contact with each other through many stages. If the equilibrium constant is equal to 1, no separation will take place, and for adjacent rare earths it is difficult to find complexes—except in special cases—that differ much from that value.

The liquid–liquid extraction process suffers from the disadvantage that for a given system only one cut is made in the rare-earth series, and if 15 pure rare earths were desired 14 cuts would have to be made. It also suffers from the disadvantage that the distance the rare-earth ions must travel in order to complete one exchange is many times that required in the ion-exchange columns. It has the advantages, however, that more concentrated solutions can be used and that the process is more economical for handling large quantities of materials. So far, it has found application mainly in special cases. It is used in some industries to concentrate the total rare earths where their abundance in the original materials is low. It is also used for separating certain elements, such as lanthanum, cerium, europium, and yttrium, with which favourable equilibrium constants are found. This is the case for cerium and europium, because they can be extracted in their tetravalent and divalent forms, respectively. Yttrium is not a lanthanide, and its position in the rare-earth series can be changed by using different organic solvents or complexing agents. First, a complex is used that separates the yttrium and heavy lanthanides from the light lanthanides, and then a different system is used whereby the yttrium is shifted with respect to the lanthanide series, so that it can be separated from the heavy lanthanides.

The liquid–liquid extraction system has not been successful in separating the adjacent heavy rare earths in the quantities desired. If ultrahigh-purity rare earths are required, it is common practice—even in those cases where liquid–liquid extraction methods have been used—to place the somewhat impure rare earth on an ion-exchange column and to use the displacement method for further purification.

Preparation of pure metals. *Early metal-reduction methods.* It is relatively easy to reduce anhydrous halides of the rare earths to metals. What is difficult, however, is to reduce them to high-purity metals in ingot form. The rare-earth metals have a great affinity for the nonmetallic elements—hydrogen, boron, carbon, nitrogen, oxygen, silicon, sulfur, phosphorus, chlorine, and bromine—and form very stable compounds with them. If a small amount of rare-earth metal is added to most other metals containing these elements present as impurities, it reacts with the impurities and removes them by gathering them together in nodules or transferring them to the slag phase. There has been a steady market for misch metal, a mixed rare-earth alloy, since Auer von Welsbach's time. A small addition of this alloy greatly improves the mechanical properties of many impure metals or alloys.

Also, hot rare-earth turnings (chips or curls from machining) can be used to produce extremely pure helium, neon, and argon by removing hydrogen, oxygen, nitrogen, carbon dioxide, and hydrocarbon vapours. As is often the case with the rare earths, however, other—and cheaper—materials perform this function equally well, and for this reason the rare-earth elements are seldom used for this purpose.

Finally, molten rare-earth metals dissolve almost all other metals and react with most compounds. They come close to being the hypothetical universal solvent of the ancients. The molten metal attacks any crucible in which it is melted, and the final product generally is a rare-earth-rich alloy of the crucible elements.

The metallothermic process

Mosander, in 1826, was the first to reduce a rare earth to a metal. He used a metallothermic reaction (heating with active metals) to reduce anhydrous chlorides made from his ceria with metallic sodium or potassium. His yields were low, 26 percent, and the metal existed as small nuggets in a solid slag, from which they could be separated only with difficulty. The metal was very impure; it contained considerable amounts of sodium or potassium and iron and other crucible materials. It also contained considerable amounts of hydrogen, oxygen, nitrogen, and carbon, as well as a mixture of the ceria group of rare earths.

During the next hundred years, as the individual rare earths were discovered and separated, a number of scientists reduced many of the lighter rare earths to the metallic form using the metallothermic process—but sometimes varying it by substituting calcium, magnesium, and aluminum as the reductants and anhydrous fluoride salts as the reactants. Because of the scarcity of pure rare earths, however, as well as the difficulty in finding suitable crucible materials and the poor equipment for keeping out atmospheric gases, the metals were still so impure that no extensive studies could be made of their properties.

In 1935, samples of the purest rare-earth chlorides available were reduced to metals at relatively low temperatures in glass capsules with potassium vapour. This process gave free metals in the form of fine powder imbedded in potassium chloride; no attempt was made to separate the metal from the potassium chloride, because only such properties as crystal structure, density, and magnetic susceptibility were under investigation. Potassium chloride acted as an internal standard in the X-ray investigations, and magnetic susceptibilities could be corrected for the potassium chloride present. Although these metals were not really pure by modern standards—they contained appreciable amounts of potassium and rare-earth impurities—they yielded values for the lattice constants and densities of most of the rare-earth metals that lie within 1 percent of the best modern values.

Electrolytic processes

In 1875, the first successful preparation of rare-earth metals by an electrolytic process was reported. About five grams each of cerium, lanthanum, and didymium (neodymium and praseodymium) in compact form were prepared by electrolyzing the fused chlorides covered with layers of ammonium chloride. The electrolytic technique was later improved, and, in the period 1902–05, misch metal, cerium, lanthanum, praseodymium, neodymium, and samarium were prepared. In 1906, Auer von Welsbach started the commercial production of lighter flints, for which the misch metal was electrolytically reduced. In the years 1923–26, several improvements in the cell designs were made, and somewhat purer samples of lanthanum, cerium, and neodymium were prepared, along with some yttrium, although most of the latter metal deposited as power.

The electrolytic process suffers from much the same difficulties as the metallothermic methods. It is difficult to find electrodes and cell materials that will stand up to molten rare earths and at the same time not introduce impurities into the ingot. It is also difficult to design cells that exclude all the atmospheric elements. The method works best for the low-melting rare earths, with which the cells can be kept sufficiently hot so that a molten pool of the metal forms in the bottom of the cell. With the higher melting rare earths, only powdered metal is formed, and it is difficult to separate it from the electrolyte in a pure form.

In 1931, a cell especially designed for producing much purer metals and also capable of reducing the halides of the heavier rare earths was employed to produce a quantity of cerium that contained only a small percentage of impurities and, somewhat later, the same apparatus was used to produce a number of other rare-earth metals, including europium, gadolinium, and yttrium.

By 1939 most of the rare-earth metals had been made in fair purity, and a number of their properties, such as magnetic behaviour, melting point, density, crystal structure, and chemical reactivity had been studied. All of these metals contained small amounts of metallic impurities and unknown amounts of nonmetallic impurities. Most of these impurities were not reported because analytical methods to determine them had not been developed at that time. Almost no work had been done on the properties of the rare-earth alloys except for those of cerium and lanthanum.

Importance of impurities

Modern techniques for producing ultrapure rare-earth metals. As purer rare-earth metals are produced, it is increasingly clear that many of their properties are extremely sensitive to small amounts of impurities. This phenomenon is particularly true with regard to magnetic and to nonmetallic impurities. For many industrial uses extreme purity is not required—nor even desired—since less pure metals can be produced much more cheaply. On the other hand, the presence of impurities can be critical in metal produced for research purposes, especially when experimental properties are being compared with predicted values, or in metal to be incorporated into solid-state devices. The processes described below are those used in making research-grade metal. If less pure metal is satisfactory, many of the steps described can be omitted, and the process can be terminated at the point where the desired purity is attained.

One especially favoured reduction process utilizes metallic calcium (Ca) and the rare-earth fluoride (MF_3). The reaction is as in the following equation:

$$3Ca + 2MF_3 \rightleftharpoons 2M + 3CaF_2.$$

Other metallothermic processes, however, can also be used, such as lithium (Li) metal and the rare-earth chloride (MCl_3):

$$3Li + MCl_3 \rightleftharpoons M + 3LiCl.$$

Variations of these methods, using lithium, sodium, potassium, or calcium as the reducing agent and any halide of the rare earth for the reactant, also are possible. For these alternative processes to succeed, however, the metal must be separable from the slag cleanly without introducing impurities, and all sources of contamination must have been eliminated.

The problem of obtaining sufficient quantities of highly pure individual rare-earth oxides has been solved by the development of the displacement-band method of separating rare earths on ion-exchange columns described above. If the oxides are obtained from the middle third of the pure rare-earth band, the total of other rare-earth impurities in the metals obtained from them does not exceed 10 parts per million. Fractions taken closer to the band edges contain somewhat larger proportions of such impurities. If the same equipment is used to prepare the different raw materials and to make the metals of a number of different rare earths, great care must be taken to prevent cross contamination of the rare earths.

Limiting contamination

Contamination from the crucible cannot be eliminated entirely. Tungsten and tantalum make the best crucibles: they are little attacked by molten rare-earth metals at temperatures below 1,000° C, and the crucible material introduced into the rare-earth metals at higher temperatures, if harmful, can be removed by special techniques. Both tungsten and tantalum are available commercially in the form of both crucibles and thin sheets: to prepare them for use, these materials are thoroughly cleaned and baked in a high vacuum to remove impurities that may be adsorbed on their surfaces.

Introduction of impurities from the atmosphere can be largely eliminated by carrying out all operations in an environment of purified helium and by the use of modern high-vacuum ion pumps instead of oil pumps.

Finally, if ultrapure metals are to be obtained, the raw materials from which they are made must also be ultrapure or their impurities will end up in the rare-earth ingot. Commercial calcium is doubly distilled at low pressure in an atmosphere of pure argon or helium to remove iron and various nonmetals that it contains, and thereafter is rigorously protected from carbon dioxide and water vapour. Anhydrous rare-earth halides form oxyhalides upon contact with water vapour; therefore, the preparation of an anhydrous fluoride is carried out in two steps. The first is the passage of dry hydrogen fluoride over the powdered oxide, immediately sweeping away any water formed; and the second is the passage of the pure, dry hydrogen fluoride over the molten fluorides. If this is done, the oxygen content of the fluoride can be kept below ten parts per million.

Groups for standardized operation

There are enough differences in the properties of the 17 rare-earth metals that the same reduction process does not work equally well for all of them. If metal containing less than 0.01 weight percent of impurities is desired, each element has to be treated somewhat differently. Many of the operations are the same for all reductions, and if the metals are divided into five groups, standardized operations can be applied for all metals in a group. The groups are as follows: Group I consists of those metals that have low melting points and high boiling points—lanthanum, cerium, praseodymium, and neodymium. Group II consists of those metals having high melting points and high boiling points—gadolinium, terbium, scandium, yttrium, and lutetium. Group III consists of those metals having high melting points, low boiling points, and, in addition, an appreciable vapour pressure at the melting point—dysprosium, holmium, erbium, and scandium. Group IV consists of those metals that have low boiling points—samarium, europium, ytterbium, and thulium. Group V, consisting only of promethium, would be included in Group II, except that serious difficulties result from the intense radioactivity of the metal. All operations with it must be carried out by remote controls, and this is usually done at special installations.

The calcium-reduction process works well for the metals of Groups I, II, III, and V, but not at all well for Group IV. Most of the trifluorides (SmF_3, EuF_3, YbF_3) of this group are reduced only to divalent fluorides, even when a large excess of calcium is used; the resulting material resembles, but is not, the metal. Thulium trifluoride is reduced to the metal, but the high vapour pressure of molten metallic thulium causes difficulty.

The standard procedure for producing the metal by calcium reduction is to load a tantalum container with enough rare-earth fluoride to yield a metal billet weighing about 300 grams. About 10 percent excess calcium is added to drive the reaction to completion. The crucible is then placed in a furnace with a helium atmosphere and heated above the melting point of the rare-earth metal or of the slag—whichever is greater—and held at that temperature until the reaction is complete and the metallic and slag layers have separated because of differences in their densities. After cooling to room temperature, the crucible is taken out of the furnace in the dry box, cut in two at the metal–slag interface, and all slag is knocked off the metal. Usually a bright metal surface can be obtained. The metal ingot, however, contains small amounts of calcium and rare-earth fluoride as impurities. Group I and Group II metals are then put in another tantalum crucible and replaced in the furnace for the boiling-off process. This time a high vacuum is used, and the metal is heated to about 1,400° to 1,500° C and held there for some time, so that any volatile impurities, particularly calcium and rare-earth fluoride, evaporate. At these temperatures considerable amounts (1 to 3 percent) of tantalum dissolve in the molten metal. The furnace temperature is then slowly lowered until it is just above the melting point of the pure metal, at which temperature it is held for a few minutes to allow most of the tantalum to precipitate onto the walls or sink to the bottom of the crucible. (For Group I metals, the solubility of tantalum is about 50 parts per million or less at the melting point.) The ingot is then removed from the cooled furnace in the dry box, and the tantalum crucible and precipitates are machined off. The resulting ingot usually contains less than 0.01 percent total impurities.

Removal of dissolved tantalum

The Group II metals, because of their higher melting points, still contain some tantalum as an impurity when they solidify; usually this tantalum appears as a second phase, showing up as black dots in the metallographic pictures of the metal. It is possible, however, to purify these metals further by distillation from a tantalum still. The still consists of a short tantalum crucible located in a high-vacuum furnace. Affixed to this crucible is an inverted crucible, which is out of the heating zone of the furnace, its upper part being 400° to 500° C cooler because of radiation losses. The rare-earth metal can then be slowly sublimed (changed from a solid into a gas without passing through the liquid state) and resolidified in the inverted crucible. Because volatile impurities usually have different boiling points and heats of sublimation, it is possible (by choosing the right temperature) to sublime the metal in such a manner that the impurities can be separated from the metal. The nonvolatile tantalum remains behind in the still. Finally, the condenser, with its rather porous crystalline mass of rare-earth metal, is removed from the furnace, and the tantalum (or tungsten) is machined off in the dry box. The porous mass is then arc-melted into a billet on a water-cooled copper hearth under an inert atmosphere.

The Group III elements cannot be held at their melting points for long: because of their volatility a considerable quantity of the metal is lost. The boiling process therefore is omitted, but sublimation or volatilization works well, and—by the right choice of temperature in the still—both the volatile and inert impurities can be eliminated.

For the Group IV metals, only a distillation process is used. The pure oxide of the rare earth is dissolved in acid and reprecipitated using ultrapure chemicals to remove traces of calcium and magnesium often introduced from the water and chemicals used in the ion-exchange process. The precipitate is again converted to the oxide and placed in the still. Pure metallic lanthanum, cerium, or misch metal, which has been subjected to the boiling-off process to remove volatile impurities, is added in excess.

The reaction of europium oxide with lanthanum metal ($Eu_2O_3 + 2La \rightarrow La_2O_3 + 2Eu\uparrow$) takes place when the mixture is heated. Because the vapour pressure of europium is millions of times greater than that of lanthanum or of the oxide of either element, the metal distills away from the oxides, and the reaction goes practically to completion. If ultrahigh-purity metal is desired, a second distillation is performed.

Problems with thulium

Thulium poses special problems. Its melting point is so high that the molten metal acquires a considerable amount of impurity from the crucible. On the other hand, it has such a high vapour pressure at the melting point that it is practically impossible to melt it without losing much of the metal. Thulium is never melted, therefore, but is sublimed to the condenser, on which it forms solid crystals but not compact metal. If a solid bar is desired, the porous metal can be pressed into a tantalum tube and reduced to about half its diameter. The tantalum covering can then be machined off and a bar of compact metal obtained.

Major industrial uses. The properties of the 17 rare-earth elements in the form of their metals, alloys, or compounds—or some combination thereof—are so varied as to make them valuable for many industrial uses. Many other somewhat less costly materials, however, often will perform just as well; and when this is the case, the rare-earth elements are seldom used for these purposes. Only when their properties are unique is the extra cost justified industrially.

Use as catalysts

Millions of tons of rare earths have been used annually in the United States to produce catalysts for the cracking of crude petroleum. The natural mixture of rare earths obtained from the minerals accounted for about 20 percent of that total, and the remaining 80 percent was made up of special mixtures of lanthanum, praseodymium, neodymium, and samarium. Rare-earth catalysts have been repeatedly recommended for use in numerous organic reactions, including the hydrogenation of ketones to form secondary alcohols, the hydrogenation of olefins to form alkanes, the dehydrogenation of alcohols and butanes, and the formation of polyesters. The extent to which these catalysts are used in industry seldom is made public, but there is no doubt that the rare earths show marked catalytic properties.

Another substantial use of rare-earth oxides is in the glass industry. Cerium oxide has been found to be a more rapid polishing agent for glass than rouge, and several million pounds a year are consumed in the polishing of lenses for cameras, binoculars, and eyeglasses, as well as in polishing mirrors and television faceplates. Glasses containing lanthanum oxide have very high refractive indexes and low dispersions. Such glasses are used in complex lenses for cameras, binoculars, and military instruments—for the purpose of correcting spherical and chromatic aberrations. Rare-earth oxides often are added to glass melts in order to produce special glasses. Neodymium is added to some glasses to counteract the yellowish tint caused by iron impurities. Very pure neodymium oxide, when added in sufficient quantities (1–5 percent), gives a beautiful purple glass. Praseodymium and neodymium are added to glass to make welders' and glassblowers' goggles, that absorb the bright-yellow light from the sodium flame. The same combination is sometimes added to the glass used in television faceplates to decrease the glare from outside light sources. A beautiful yellow ceramic stain results from the addition of about 3 percent praseodymium oxide to zirconium oxide. Cerium oxide increases the opacity of white porcelain enamels.

Use in metallurgy

The metallurgical industry is another heavy user of rare earths. Small amounts of misch metal and cerium have long been added to other metals or alloys to remove their nonmetallic impurities. Misch metal added to cast iron makes a more malleable nodular iron. Added to some steels, it makes them less brittle. The addition of misch metal to certain alloys has been reported to increase the tensile strength and improve the hot workability and the high-temperature oxidation resistance. The rare earths are particularly effective in iron–chromium and iron–chromium–nickel alloys to improve a number of their properties, especially their resistance to corrosion and oxidation. Yttrium metal is said to work even better than misch metal in removing impurities from certain materials. The flints of cigarette lighters are an alloy of misch metal and iron.

The addition of misch metal or pure rare-earth elements to magnesium increases its high-temperature strength and its creep resistance—that is, resistance to slow deformation under prolonged use. This alloy also makes better castings if small amounts of zirconium or other metals are added, and such alloys are used in jet-engine and precision castings. The addition of small amounts of rare-earth elements to aluminum has also been reported to give better castings.

Use in television screens

By far the heaviest user of ultrapure separated rare earths is the television industry. It has been found that if a small amount of europium oxide (Eu_2O_3) is added to yttrium oxide (Y_2O_3), it gives a brilliant-red phosphor. Colour television screens utilize red, green, and blue phosphors. In the past, a zinc–cadmium sulfide was used as the red phosphor, but it was not completely satisfactory because its fluorescent band was too wide, and it could not be made to fluoresce as intensely as the other phosphors. The Y_2O_3–Eu_2O_3 phosphor corrected these disadvantages and made possible much brighter and more natural coloured pictures. This use has been growing in many countries. Many of the early rare-earth screens used europium–yttrium orthovanadate phosphor, but the industry is shifting heavily toward the oxide phosphor. Some television companies have substituted gadolinium oxide for the yttrium oxide. The rare-earth phosphors are also finding use in mercury-arc lights, which are used for sporting events and special street lighting. Instead of the unhealthy-looking blue light of the mercury arc, the phosphors give an intense white radiation similar to daylight. Considerable amounts of mixed rare-earth fluorides are used to make cored carbon rods, which are used as arcs in searchlights and in some of the lights used by the motion-picture industry.

Yttrium-iron garnets are synthetic high-melting silicates that can be fabricated into special shapes for use as microwave filters in the communications industry. Yttrium-aluminum garnets also are being produced at an increasing rate for use both in electronics and as gemstones. Both of these synthetic minerals have much use in the jewelry business. These garnets have a high refractive index and a hardness approaching that of diamond. In a solid crystal form they are amazingly transparent, and they are being cut into imitation diamonds.

Another significant industrial application of rare earths is in the manufacture of strong permanent magnets. Alloys of cobalt with rare earths, such as cobalt–samarium, produce permanent magnets that are far superior to most of the varieties now on the market. Another relatively recent development is the use of a barium phosphate–europium phosphor in a sensitive X-ray film that forms satisfactory images with only half the exposure.

Nuclear applications

Europium, gadolinium, and dysprosium have large capture cross sections for thermal neutrons—that is, they absorb large numbers of neutrons per unit of area exposed. These elements, therefore, are incorporated into control rods used to regulate the operation of nuclear reactors or to shut them down should they get out of control. In addition, rare-earth elements are used as burnable neutron absorbers to keep the reactivity of the reactor more nearly constant. As uranium undergoes fission, it produces some fission products that absorb neutrons and tend to slow down the nuclear reaction. If the right amounts of rare-earth elements are present, they burn out at about the same rate that other absorbers are formed.

Yttrium dihydride is used as a moderator in reactors to slow down neutrons. Certain rare earths are also used in shielding materials because of their high nuclear cross sections. Scandium metal is used as a neutron filter that allows neutrons only of a certain energy (two kiloelectron volts) to pass through.

Complexes of europium, praseodymium, or ytterbium with derivatives of camphor are useful reagents for analysis of optically-active organic compounds, which often are obtained as mixtures containing unknown proportions of two components that differ only in that their molecular structures are mirror images of each other. Determination of these proportions can be very difficult, but the rare-earth complexes provide asymmetric environments in which each component absorbs electromagnetic radiation of a particular frequency in the presence of a strong magnetic field. Proportions then can be determined by measuring the intensities of the separate absorptions.

Low toxicities of the rare earths

The rare earths have low toxicities and can be handled safely with ordinary care. Solutions injected into the peritoneum will cause hyperglycemia (excess of sugar in the blood), decreased blood pressure, spleen degeneration, and fatty liver. If solutions are injected into muscle about 75 percent of the rare-earth element remains at the site, the remainder going to the liver and skeleton. When taken orally, only a small percentage of a rare-earth element is absorbed into the body. Organically complexed ions are somewhat more toxic than solids or inorganic solutions. As is true for most chemicals, dust and vapours should not be inhaled, nor should they be ingested. Solutions splashed into the eyes should be washed out, and splinters of metal should be removed.

When handling rare-earth ores or minerals, dust should be avoided because many minerals contain other toxic elements, such as beryllium, thorium, and uranium. Finely divided rare-earth metals can ignite spontaneously, somewhat as magnesium does. (F.H.S./Ed.)

Actinide elements

The actinide elements constitute a group of 15 consecutive elements in the periodic table from actinium, symbol Ac, with atomic number 89, to lawrencium, symbol Lr, and atomic number 103. As a group they are significant largely because of their radioactivity. Although several members of the group, including uranium (the most familiar), occur naturally, most are man-made. Both uranium and plutonium have been used in atomic bombs for their explosive power and currently are being employed in nuclear plants for the production of electrical power.

GENERAL COMPARISONS

The actinide elements follow one another in the seventh series of the periodic table (see Figure 10). Each has 86 electrons arranged as in the atoms of the noble gas radon (which precedes actinium by several spaces in the table), with three more electrons that may be positioned in the *6d* and *7s* orbitals (the seventh shell is outermost), and with additional electrons packing into inside orbitals. Specifically, the series is formed by the insertion of one more electron for each successive new element into an underlying *5f* orbital. The valence electrons, however, are found mainly in the *6d* and *7s* orbitals. Thus, the chief difference among the atoms of the elements of the series is the single additional electron deep within the electron cloud; but the fact is that, because of its position in the 5th shell, this distinguishing electron actually affects the chemical properties of the actinides only in a relatively minor way; *5f* electrons are not involved in the formation of chemical bonds with other atoms.

As is usual with the elements of any group, there are a number of exceptions to these generalities, particularly in the lower members of the series; but for most of these elements, the concept of a series of nearly identical actinide elements is a useful guide for predicting their chemical and physical properties.

Like all elements, each actinide has its own particular atomic number, equal to the number of protons in the nucleus and, consequently, to the number of electrons; at the same time the atoms of an element are capable of existing in a number of forms (isotopes), each of which has a different number of neutrons in its nucleus and hence a different atomic mass. Although isotopes of a given element behave alike chemically, they may have different stabilities in relation to radioactive decay, which is a property of the nucleus. No element beyond bismuth in the periodic classification—*i.e.,* no element that has an atomic number greater than 83—has any stable isotopes; radioactive isotopes of every element in the table can be produced in the laboratory. The actinides are unusual in forming a series of 15 elements having no stable isotopes; every actinide isotope undergoes radioactive decay, and, as a result, only a few of the lighter, more stable members of the series (such as thorium and uranium) are found in nature. The half-life, or the precise time required for one-half of any amount of a particular isotope to disappear due to radioactive decay, is a measure of the stability of that isotope. The naturally occurring isotopes in the actinide series have long half-lives, of the order of billions of years.

PRACTICAL APPLICATIONS OF THE ACTINIDES

Nuclear fission

The practical significance of the actinides arises from the fissionability, or potential for splitting, of certain of their isotopes. When an atomic nucleus breaks apart, or undergoes fission, a far more disruptive process than ordinary radioactive decay, enormous amounts of energy are liberated. This energy can be allowed to generate an atomic explosion, or it can be controlled and used as a fuel to generate heat for the production of electrical power. Nuclear processes for power production give off no smoke, smog, or noxious gases, as conventional coal- or gas-fueled plants do. Nuclear power plants, however, do yield waste heat that may be considered as thermal pollution, and they also yield useless and dangerous radioactive wastes which, although they are pollutants, may be less undesirable than those from fossil-fuel generators. For this and other reasons, such as economy of operation, there is a potential for an enormous electrical energy production inherent in nuclear energy generating technologies, and, since the actinide elements are the only known fissionable materials, the practical impact of their availability is great. The isotope of uranium with the atomic number 92 and mass 235, written as uranium-235 or, in chemical symbols, as ^{235}U, is present to the extent of only about 0.7 percent in ordinary uranium, but it is a necessary fissionable material in the operation of a nuclear reactor using natural uranium. Other fissionable isotopes of great importance are uranium-233, plutonium-239, and plutonium-241.

Fissionable plutonium isotopes are formed as by-products of fission in reactors using uranium; when neutrons are added to uranium-238, which is not itself fissionable, it is converted to the fissionable isotope plutonium-239. Thorium, also, is potentially of great economic value, because one of its isotopes, thorium-232, can be converted into the fissionable isotope uranium-233 in a nuclear breeder reactor (*i.e.,* one that produces more fissionable material than it consumes), thus increasing by many times available supplies of fissionable materials. Since thorium is about three times more plentiful than uranium, the potential use of this element to produce nuclear energy is significant.

The heavier actinides, those beyond plutonium in the periodic table, are of interest principally to research scientists, though they have some practical uses as sources of thermoelectric heat and neutrons. They are employed to some extent in cancer therapy.

Actinium, thorium, protactinium, and uranium are the only actinide elements found in nature to any significant extent. The remaining actinide elements, commonly called the transuranium elements, are all man-made by bombarding naturally occurring actinides with neutrons or with heavy ions (charged particles) in particle accelerators (such as cyclotrons). The actinides beyond uranium do not occur in nature (except, in some cases, in trace amounts), since the stability of their isotopes decreases with increase in atomic number and whatever quantities may be produced decay too fast to accumulate. The half-life of uranium-238, the most stable uranium isotope, is 4.5×10^9 years. Plutonium-239 has a half-life of 24,400 years and is produced in reactors in ton amounts; but nobelium and lawrencium, elements 102 and 103, with half-lives of seconds, are produced a few atoms at a time. The first of these synthetic actinide elements to be discovered (1940) was neptunium, symbol Np, atomic number 93, which was prepared by bombardment of uranium metal with neutrons (see below *Transuranium elements*).

PROPERTIES OF THE GROUP

Similar lanthanide and actinide series

Chemical properties. The chemistry of any element can be understood best in terms of atomic structure and its effect on the formation of chemical bonds. In the actinide series, just as in the lanthanides, added electrons (with increasing atomic number) go into internal *f* orbitals, where they are partially buried and consequently not chemically active. These two series occur in Group IIIb of the periodic table; because the outer, or valence, electrons of these elements are much the same, the chemical properties of the elements in the two series tend to resemble one another closely. A great deal is known about the lanthanides, all but one of which occur in nature as stable isotopes and, therefore, predictions about the chemistry of the actinides, some of which can be prepared only in minute quantities, can be made with some success by comparing their electron structures with those of the lanthanides. In the lanthanide series of elements, as indicated above, each added electron goes into the *f* orbital of the fourth shell; these orbitals are designated as 4*f.* In the actinide elements, the added electrons also go into an *f* orbital in a similar manner, but in the fifth shell, instead. Electrons with larger quantum numbers generally are farther from the nucleus than those with smaller quantum numbers, and are therefore usually less strongly held by it. As expected, then, electrons in the 5*f* orbitals, being farther from the nucleus, are much less tightly bound than are those in 4*f* orbitals and, in fact, sometimes are active enough to take part in chemical reactions. The result is that the actinide elements, in which the 5*f* orbitals are being filled, have more variable valences (number of electrons available for chemical bonds) than do the lanthanides, in which the 4*f* orbitals are being filled.

The similarities between many lanthanide and actinide compounds are striking and offer a useful comparison. Under certain conditions, for example, actinium, americium, curium, and berkelium metals have the same crystal structure, as do many of the lanthanides. It is not yet known whether californium or einsteinium, the only other actinide elements with sufficiently stable isotopes for chemical work, have the same structure. Several of the lighter actinide elements from thorium through plutonium have different and unusual metallic structures, presumably due to the mixing of 5*f* and 6*d* orbitals in their atoms, some electrons entering unfilled 6*d* orbitals rather than the expected 5*f* orbitals.

The +2 oxidation state

The actinides generally show multiple oxidation states. Compounds of americium and californium with an oxidation state of +2 are known; there are reasons for expecting the existence of this state in some of the elements heavier than californium. For example, spectroscopic evidence for einsteinium(II) in the presence of the fluoride ion has been obtained. Divalent actinides (that is, actinides in the +2 oxidation state) form compounds with nearly the same properties as those of the divalent lanthanides and, accordingly, iodides, bromides, and chlorides of divalent americium and californium have been found to be stable. If X symbolizes the nonactinide and M symbolizes the actinide element, then the general formula for these compounds would be one atom of actinide and two atoms of nonactinide: MX_2.

The +3 oxidation state

Great similarities in chemical behaviour are found in the actinides of oxidation state +3, from actinium to einsteinium; furthermore, they are much like the lanthanides of the same oxidation state. The crystal types and many physical properties of these trivalent actinides are dependent more on the size of the +3 ion (an atom that has given up three electrons and has become an ion with three positive charges, symbolized as Ac^{+3}, etc.) of the particular element that is involved. For instance, the solubility of the trifluorides formed by actinides with a +3 state (thorium and protactinium have no such state) is exceedingly low. The crystal-structure type for the actinide trifluorides is the same as that of lanthanum trifluoride, and, since the radius of the ion is a regular function of the atomic number, the circumstance allows extrapolation from the lanthanum compound to the actinide compound and interpolation between known compounds in the series to determine missing values. The hydroxides, phosphates, oxalates, and alkali double sulfates of the actinides are also insoluble, with many of each having identical crystal structures, or being isostructural. The chlorides, bromides, and iodides (*i.e.,* the halides) of the actinides are, for the most part, isostructural for any one halide, and the structure type can be predicted from a knowledge of the ionic radius. The solubility of these halides in water is generally great. The +3 oxides of actinides are also isostructural, with the general formula M_2O_3, in which M is any of the actinide elements; they form cubic (or hexagonal) crystals, and the sizes of the molecules are thus easily predictable. Generally, then, the chemistry of the actinides in the +3 oxidation state is similar, with the differences mainly due to ionic size. As a consequence of these similarities, separations of the elements and of their components are frequently difficult, necessitating the use of methods in which very slight physical differences of the atoms or ions serve to separate the chemically almost identical materials. Two methods are ion-exchange reactions, in which differences in ions and bonding are used to effect separation and solvent extraction, in which specific solvents are used to dissolve and withdraw from the mixture the desired element or its compound.

The +4 oxidation state

Actinides in the +4 oxidation state also are much alike (and also resemble the +4 lanthanides). The +4 actinides, thorium, protactinium, uranium, neptunium, plutonium, berkelium, and, to a lesser extent, americium and curium, are sufficiently stable to undergo chemical reactions in water solutions. Crystallized compounds in the +4 state exist for thorium, protactinium, uranium, neptunium, plutonium, americium, curium, berkelium, and californium. The oxides and many complex fluorides are known for all of these elements. The dioxides are all isostructural, as are the tetrafluorides. Most of these actinide compounds can be prepared in a dry state by igniting the metal itself, or one of its other compounds, in an atmosphere of oxygen or of fluorine. Some tetrachlorides, bromides, and iodides are known for thorium, uranium, and neptunium. The ease with which they can be formed decreases with increasing atomic number. Berkelium(IV) appears to be sufficiently stable to allow the preparation of the tetrachloride. Hydroxides of a number of these elements in the +4 state also are known; they are of very low solubility, as are the fluorides, oxalates, and phosphates. Again, many physical properties of the tetrafluorides are influenced more by ionic size than by atomic number, and isostructurality of these actinide and lanthanide compounds is the rule rather than the exception.

The similarities exhibited by the lanthanide and actinide compounds in the +3 and +4 oxidation states, as well as in some cases by the free elements, can be very useful. A great many individual differences, however, do arise. These are due, in part, to mixing of the orbitals (some electrons moving into *d* rather than *f* orbitals) and, in part, to the relative degrees of binding of the *f* electrons.

Special properties. *Chemical.* The metals thorium, protactinium, uranium, neptunium, and plutonium are for the most part different from one another. Uranium, neptunium, and plutonium have extremely dense metallic forms. Neptunium, for example, with a density of 20.45 grams per cubic centimetre when crystallized into the orthorhombic crystal form, is one of the densest metals known to man. A possible explanation for the fact that these metals show a number of different crystal forms is that the electrons in the 5*f* orbitals mix with those in the 6*d* orbitals and consequently form a number of hybrid electronic states of nearly the same energy. Beginning with americium, however, the electron energy levels seem to be sufficiently separated so that mixing does not occur.

The +5, +6, +7 oxidation states

The other main differences shown by the actinides is that some possess the +5, +6, and +7 oxidation states (no lanthanide element exceeds the +4 state). It appears that the 5*f* electrons of the actinides, being far enough from the positively charged nucleus, permit increasingly easier removal and consequent formation of higher oxidation states. The element protactinium shows the +5 state; uranium, neptunium, and americium exhibit the +5 and +6 states; only neptunium and plutonium have the +7 state.

There are two types of chemical reactions for the +5 and +6 states. If M symbolizes any actinide, and O, as usual, symbolizes oxygen, then the ions found both in aqueous (water) solution and in solids prepared from solution are represented by the general formulas MO_2^+ (meaning a molecule consisting of one atom of M with two of oxygen, the whole having a single positive charge) and MO_2^{2+}. In nonaqueous solution, and in solids prepared from them, compounds of M that do not contain oxygen are known. With the halogens (X being a general designation for a halogen—fluorine, chlorine, bromine, or iodine), compounds are known that can be represented as MX_5 (meaning a molecule consisting of one atom of an actinide with five atoms of a halide) and MX_6, as well as complexes of the type having the molecular formulas MX_6^-, MX_7^{2-}, and MX_8^{3-} for the +5 states, and MX_7^- and MX_8^{2-} for the +6 states. Neptunium(VII) and plutonium(VII) have been prepared in basic solution, and certain oxygenated ions (of the type represented by MO_5^{3-}) as well as a few solid compounds, have been identified with the same oxidation state. Complex oxides with alkali metals in which these two elements have the +7 state also have been prepared.

Physiological. All the actinide elements are heavy metals and, as such, are toxic, just as lead is toxic—relatively large amounts ingested over a long period cause serious illness. But, with the exception of the long-lived thorium and uranium isotopes, the real danger with the actinide elements lies in the radioactive properties of these elements. They are emitters of tissue-destroying and cancer-producing rays (alpha, beta, or gamma; see ATOMS: *Radioactivity*). Furthermore, the chemistry of many of these elements is such that, once ingested, they tend to remain in the body almost indefinitely. Several, such as plutonium and americium, migrate to the bone marrow, where their radiation interferes with the production of red blood cells. As a consequence, workers using these elements are required to take elaborate precautions to prevent ingestion. Less than one-millionth of a gram can be fatal.

Radioactive decay series

Nuclear. All of the actinide elements are unstable toward radioactive decay; the reason that actinium, thorium, protactinium, and uranium are found in nature at all is because some of their isotopes are unusually stable and others are being formed constantly by decay of the long-lived isotopes. It is convenient to divide the naturally occurring isotopes into families based on the relationships of their atomic masses to each other. The mass numbers of all isotopes of the so-called thorium series, for instance, turn out to be multiples of four, and the series is known as the 4n series. In the uranium series, masses have been shown to be such that they are represented by 4n + 2; and in the actinium series, by 4n + 3. Not found in nature to any significant extent, but which can be artificially produced by man, is the neptunium, or 4n + 1 series, named for its longest-lived member, neptunium-237.

In the various radioactive decay processes, several "rays" are emitted, the term ray being a holdover from the time when all of these emissions were thought to be rays. They are: (1) electrons, called beta particles to indicate their origin in radioactive decay, and designated as negative beta, or $\beta-$, particles; (2) helium nuclei, called alpha particles and designated as α particles, or as helium with a plus two charge, He^{+2}; (3) gamma rays, which are electromagnetic waves of very high frequency, designated as γ rays; (4) positrons, which are positively charged electrons and are designated as positive beta, or $\beta+$, particles. Finally, an orbital electron in a radioactive atom may be captured by the nucleus and taken into it. This radioactive event is called K-capture. Except for the emission of gamma rays, each of these processes leads to an isotope of a different element; that is, to a substance with a different atomic number. The emission of an alpha particle leads to a change in atomic weight as well, because this emission has significant mass. The changes occur in sequence, each decay process leading to still another unstable element, until a region of stable isotopes of the elements lead and bismuth is reached.

The most important nuclear reactions, however, involve the capture of neutrons by an actinide nucleus, followed by splitting, or fission, of that nucleus into two unequal parts, with the liberation of enormous quantities of energy plus two or more extra neutrons. Nuclear reactors and atomic bombs depend upon the chain reaction set up by this process: the resulting neutrons react further, inducing more fission reactions, which produce more neutrons, which lead to still more fission reactions, with the result that, without control, a great deal of energy is liberated very quickly.

One gram of burning coal yields fewer than 10,000 calories of heat. The fission of one gram of uranium-235 produces 2×10^{10} calories, or about 2,000,000 times as much energy. With adequate control the energy can be released in useful form.

Fissionable isotopes

The most important actinide by far, because of its fissionability, is uranium, which has several isotopes. Natural uranium consists mostly of uranium-238, a nonfissionable isotope; uranium-235 occurs to the extent of only seven-tenths of 1 percent; and there are even smaller amounts of uranium-234.

The fissionable isotope uranium-235 is the valuable substance that causes the chain reactions with neutrons. Methods that separate isotopes by virtue of their different masses are used to eliminate the unwanted isotopes in the naturally occurring mixture of uranium ores and thus enrich natural uranium with respect to uranium-235. Such enriched uranium was used in the first atomic bomb exploded over Japan.

Two other fissionable actinide isotopes are important and undoubtedly will be more so in the future. The first is the isotope plutonium-239 prepared by neutron bombardment of uranium-238. First, uranium-239 is formed by neutron capture; this isotope then undergoes negative beta decay, giving neptunium-239, which in turn reacts by negative beta decay to give plutonium-239. Plutonium-239 thus is formed as a by-product in uranium piles in which uranium-238 is exposed to neutrons; it can be used in place of uranium-235 in atomic weapons or in reactions. The second fissionable isotope is potentially even more important, because it is produced from the element thorium, of which there are enormous reserves on Earth. When the isotope thorium-232 is bombarded by neutrons, one of which it captures, it is converted to thorium-233. This substance decays by beta emission to protactinium-233, which again emits a beta particle to give uranium-233, also a fissionable isotope of uranium. Since uranium is a relatively scarce element, development of nuclear power on a large scale is expected to deplete rapidly the uranium that can be produced economically. The use of thorium thus could extend the supplies of fissionable material by about threefold, and the use of plutonium as well could more than double them. Breeder power reactors are designed so that very few neutrons are lost through the surface or from absorption by impurities, and more fissionable material (either uranium-233 or plutonium-239) is produced than is consumed. Should such breeder reactors be made practicable, never-ending supplies of fissionable isotopes would be available.

Thermoelectric power generators

Although no other fissionable isotopes of the actinides occur in significant amount, there are practical uses of the nonfissionable isotopes due to the amounts of heat they produce by nuclear decay. For special power sources that require great dependability, thermoelectric generators using these isotopes can be considered. Alpha-emitting isotopes with radioactive half-lives in the range of several months to 100 years or more are suitable candidates. (Beta- or gamma-emitting isotopes require too much shielding to be usable.) Curium-244, which has a half-life of 17.6 years, produces 2.9 watts of heat per gram. Another isotope, plutonium-238, produces 0.57 watt per gram and has the much longer half-life of 86 years. Such power sources have been planted on the Moon to provide electrical energy for transmitting messages back to Earth. Plutonium-238 also is being considered for use as a power source for artificial hearts. A thermoelectric power generator fueled by this radioisotope already has been implanted in the human body to power a heart pacemaker, which is not expected to need replacement for at least 10 years. (L.B.As./Ed.)

Transuranium elements

The transuranium elements are so called because they all lie beyond uranium; that is to say, they are elements with atomic numbers higher than 92 in the periodic table (see Figure 10). Eighteen elements, up to and including atomic number 110, have been discovered, although confirmation of the discovery of elements after 103 remains uncertain. Eleven of them, from neptunium to lawrencium, belong to the actinide series, described above. The other seven, which have atomic numbers higher than 103, are sometimes referred to as the transactinides. All of the transuranium elements are unstable, decaying radioactively, with half-lives that range from tens of millions of years to mere fractions of a second. Since only two of the 18 have been found in nature (neptunium and plutonium) and those only in trace amounts, the synthesis of these elements through nuclear reactions has been an important source of knowledge about them. That knowledge has expanded scientific understanding of the fundamental structure of matter and makes it possible to predict the existence and basic properties of elements much heavier than element number 110—*i.e.,* the so-called superheavy elements, those with atomic numbers from 110 to around 126. Since it is predicted that such elements should have significant nuclear stability and, therefore, may exist in nature, some of them are actively being sought. Elements with atomic numbers even higher, around 164, are also predicted, but with lesser nuclear stability; and present theory suggests that the maximum atomic number will be found to lie somewhere between 170 and 210. All these still unknown elements are included in the transuranium group.

DISCOVERY OF THE FIRST TRANSURANIUM ELEMENTS

First synthesis. The first attempt to prepare a transuranium element was made in 1934 in Rome where a team of Italian physicists headed by Enrico Fermi and Emilio Segrè bombarded uranium nuclei with free neutrons. Although transuranium species may have been produced, the results of the experiment were ultimately interpreted by the German physical chemists Otto Hahn and Fritz Strassman in terms of nuclear fission. Not until 1940 was a transuranium element first positively produced and identified by two American physicists, Edwin Mattison McMillan and Philip Hauge Abelson, who, in California, exposed uranium oxide to neutrons from a cyclotron target. The resulting element was found to have an atomic number of 93. It was named neptunium.

Transformations in atomic nuclei are represented by equations that balance all the particles of matter and the energy involved before and after the reaction. The above transformation of uranium into neptunium may be written as follows:

$$^{238}_{92}\mathrm{U}\,(n,\,\gamma)\,^{239}_{92}\mathrm{U} \xrightarrow{\beta^-} {}^{239}_{93}\mathrm{Np}$$

In this equation the atomic symbol of the particular isotope reacted upon, in this case U for uranium, is given with its mass number at upper left and its atomic number at lower left: $^{238}_{92}\mathrm{U}$. The symbols within the parentheses indicate, first, the species of matter or of energy (here *n,* for neutron) reacting with the nucleus and, second, the species (here γ, for a quantum of energy called a gamma ray) emitted by the nucleus during the reaction.

Equations for nuclear transformations

The resulting isotope is shown next, with its changed mass at the upper left ($^{239}_{92}\mathrm{U}$). Together, these symbols represent the transformation of a single uranium-238 nucleus, reacting with a neutron and releasing gamma radiation, to produce the isotope uranium-239. The next event is represented by an arrow showing a spontaneous loss of a beta particle (symbolized β), an electron with very high velocity, from the nucleus of uranium-239. What has happened is that a neutron within the nucleus has been transformed into a proton, with the emission of a beta particle which carries off a single negative charge; the nucleus, therefore, now has one more positive charge than it had before the event, when its atomic number was 92. The resulting isotope, as a result, has an atomic number of 93 but, because the beta particle has negligible mass, the mass number of the nucleus has not changed. It is still 239. The nucleus resulting from these events is an isotope of the element neptunium, atomic number 93 and mass number 239. The above is called negative beta-particle decay; a nucleus may also emit a positron, or positive electron, resulting in beta-particle decay. (In another type of decay the nucleus, instead of emitting a β-particle, "captures" or absorbs one of the electrons orbiting the nucleus, thus reducing the positive charge by one.)

The usual manner of writing nuclear reactions drops the atomic number from the symbol for the element and this rule will be followed in the rest of this article.

Table 27: Selected Transuranium Elements

atomic number	element	symbol	atomic weight	discoverers¶ and date of discovery	source of first preparation	first isolation in weighable amount (half-life)
93	neptunium	Np	237*†	E.M. McMillan and P.H. Abelson at the University of California, Berkeley, 1940	irradiation of uranium with neutrons ${}^{238}\mathrm{U}\ (n,\gamma)\ {}^{239}\mathrm{U} \xrightarrow{\beta^-} {}^{239}\mathrm{Np}$	1944, ${}^{237}\mathrm{Np}$ (2.14×10^6 years) L.B. Magnusson and T.J. La-Chapelle
94	plutonium	Pu	239† 244*	G.T. Seaborg at Berkeley, 1940	bombardment of uranium with deuterons ${}^{238}\mathrm{U}\ ({}^{2}\mathrm{H}, 2n)\ {}^{238}\mathrm{Np} \xrightarrow{\beta^-} {}^{238}\mathrm{Pu}$	1942, ${}^{239}\mathrm{Pu}$ (24,400 years) B.B. Cunningham and L.B. Werner
95	americium	Am	241† 243*	G.T. Seaborg at the University of Chicago, 1944	irradiation of plutonium with neutrons ${}^{239}\mathrm{Pu}\ (2n,\gamma)\ {}^{241}\mathrm{Pu} \xrightarrow{\beta^-} {}^{241}\mathrm{Am}$	1945, ${}^{241}\mathrm{Am}$ (433 years) B.B. Cunningham
96	curium	Cm	244† 247*	G.T. Seaborg at Chicago, 1944	bombardment of plutonium with helium ions ${}^{239}\mathrm{Pu}\ ({}^{4}\mathrm{He}, n)\ {}^{242}\mathrm{Cm}$	1947, ${}^{242}\mathrm{Cm}$ (162.5 days) L.B. Werner and I. Perlman
97	berkelium	Bk	247* 249†	S.G. Thompson at Berkeley, 1949	bombardment of americium with helium ions ${}^{241}\mathrm{Am}\ ({}^{4}\mathrm{He}, 2n)\ {}^{243}\mathrm{Bk}$	1958, ${}^{249}\mathrm{Bk}$ (314 days) S.G. Thompson and B.B. Cunningham
98	californium	Cf	251* 252†	S.G. Thompson at Berkeley, 1950	bombardment of curium with helium ions ${}^{242}\mathrm{Cm}\ ({}^{4}\mathrm{He}, n)\ {}^{245}\mathrm{Cf}$	1958, ${}^{249-252}\mathrm{Cf}$ B.B. Cunningham and S.G. Thompson
99	einsteinium	Es	253† 254*	A. Ghiorso at Berkeley, 1952	irradiation of uranium with neutrons in first thermonuclear explosion ${}^{238}\mathrm{U}\ (15n,\gamma)\ {}^{253}\mathrm{U} \xrightarrow{7\beta^-} {}^{253}\mathrm{Es}$	1961, ${}^{253}\mathrm{Es}$ (20 days) B.B. Cunningham, J.C. Wallmann, L. Phillips, R.C. Gatti
100	fermium	Fm	257*†	A. Ghiorso at Berkeley, 1952	in first thermonuclear explosion ${}^{238}\mathrm{U}\ (17\ n,\gamma)\ {}^{255}\mathrm{U} \xrightarrow{8\beta^-} {}^{255}\mathrm{Fm}$	. . .
101	mendelevium	Md	258*	A. Ghiorso at Berkeley, 1955	bombardment of einsteinium with helium ions ${}^{253}\mathrm{Es}\ ({}^{4}\mathrm{He}, n)\ {}^{256}\mathrm{Md}$	. . .
102	nobelium	No	259*	A. Ghiorso at Berkeley, 1958	bombardment of curium with carbon ions ${}^{246}\mathrm{Cm}\ ({}^{12}\mathrm{C}, 4n)\ {}^{254}\mathrm{No}$	. . .
103	lawrencium	Lr	260*	A. Ghiorso at Berkeley, 1961	bombardment of californium with boron ions $\left.\begin{matrix}{}^{250}\mathrm{Cf}\\ {}^{251}\mathrm{Cf}\\ {}^{252}\mathrm{Cf}\end{matrix}\right\}\begin{matrix}+{}^{11}\mathrm{B}\\ +{}^{10}\mathrm{B}\end{matrix} \longrightarrow {}^{258}\mathrm{Lr}+\left\{\begin{matrix}2n\\ 3n\\ 4n\\ 5n\end{matrix}\right.$	. . .
104	kurchatovium‡	Ku	261*	G.N. Flerov at the Joint Nuclear Research Institute at Dubna (U.S.S.R.), 1964	bombardment of plutonium with neon ions ${}^{242}\mathrm{Pu}\ ({}^{22}\mathrm{Ne}, 4n)\ {}^{260}104$	. . .
	rutherfordium§ unnilquadium‖	Rf Unq		A. Ghiorso at Berkeley, 1969	bombardment of californium with carbon ions ${}^{249}\mathrm{Cf}\ ({}^{12}\mathrm{C}, 4n)\ {}^{257}104$ ${}^{249}\mathrm{Cf}\ ({}^{13}\mathrm{C}, 3n)\ {}^{259}104$	. . .
105	nielsbohrium‡	Ns	262*	G.N. Flerov at Dubna, 1967	bombardment of americium with neon ions ${}^{243}\mathrm{Am}\ ({}^{22}\mathrm{Ne}, 5n)\ {}^{260}105$ ${}^{243}\mathrm{Am}\ ({}^{22}\mathrm{Ne}, 4n)\ {}^{261}105$	. . .
	hahnium§ unnilpentium‖	Ha Unp		A. Ghiorso at Berkeley, 1970	bombardment of californium with nitrogen ions ${}^{249}\mathrm{Cf}\ ({}^{15}\mathrm{N}, 4n)\ {}^{260}105$	. . .
106	no name proposed unnilhexium‖	 Unh	263*	G.N. Flerov at Dubna, 1974 A. Ghiorso at Berkeley, 1974	bombardment of californium with oxygen ions ${}^{249}\mathrm{Cf}\ ({}^{18}\mathrm{O}, 4n)\ {}^{263}106$	. . .
107	no name proposed unnilseptium‖	 Uns	262*	G.N. Flerov at Dubna, 1976	bombardment of bismuth with heavy chromium ${}^{204}\mathrm{Bi}\ ({}^{54}\mathrm{Cr}, 4n)\ {}^{262}107$	. . .
108	no name proposed unniloctium‖	 Uno	. . .	Peter Armbruster at the Institute for Heavy Ion Research at Darmstadt (Ger.), 1984	bombardment of lead with iron nuclei ${}^{208}\mathrm{Pb}\ ({}^{58}\mathrm{Fe}, n)\ {}^{265}108$	. . .
109	no name proposed unnilennium‖	 Une	. . .	Peter Armbruster at Darmstadt, 1982	bombardment of bismuth with iron nuclei ${}^{209}\mathrm{Bi}\ ({}^{58}\mathrm{Fe}, n)\ {}^{266}109$	. . .
110	no name proposed ununnilium‖	 Uun	. . .	Yury Organessian at Dubna, 1987	bombardment of argon with uranium nuclei; bombardment of calcium with thorium nuclei	. . .

*Mass number of longest-lived isotope. †Mass number of more available isotope. ‡Proposed by U.S.S.R. scientists, not yet approved by Commission on Atomic Weights of International Union of Pure and Applied Chemistry. §Proposed by U.S. scientists, not yet approved by Commission on Atomic Weights of International Union of Pure and Applied Chemistry. ‖Proposed by International Union of Pure and Applied Chemistry. ¶Team leader.

For further information on discoveries of transuranium elements, see Table 27.

SYNTHESIS OF TRANSURANIUM ELEMENTS

The most abundant isotope of neptunium is neptunium-237. Neptunium-237 has a half-life of approximately 2×10^6 years and decays by the emission of alpha particles. (Alpha particles are composed of two neutrons and two protons and are actually the very stable nucleus of helium.) Neptunium-237 is formed in kilogram quantities as a by-product of the large-scale production of plutonium in nuclear reactors. This isotope is synthesized from the reactor fuels, uranium-235 and uranium-238, by the following reactions:

$${}^{238}\mathrm{U}\ (n, 2n)\ {}^{237}\mathrm{U} \xrightarrow{\beta^-} {}^{237}\mathrm{Np}$$

$${}^{235}\mathrm{U}\ (n,\gamma)\ {}^{236}\mathrm{U}\ (n,\gamma)\ {}^{237}\mathrm{U} \xrightarrow{\beta^-} {}^{237}\mathrm{Np}$$

In the first reaction, uranium-238 captures one neutron but emits two, reducing the mass number by one. Plutonium, as the isotope plutonium-239, is produced in ton quantities in nuclear reactors by the reaction:

$${}^{238}\mathrm{U}\ (n,\gamma)\ {}^{239}\mathrm{U} \xrightarrow{\beta^-} {}^{239}\mathrm{Np} \xrightarrow{\beta^-} {}^{239}\mathrm{Pu}$$

The plutonium isotope plutonium-238, which is of secondary importance, is produced in kilogram quantities by the reaction:

$$^{237}\text{Np}(n,\gamma)^{238}\text{Np} \xrightarrow{\beta^-} {}^{238}\text{Pu}$$

The production of transuranium elements by a process of continuous, intensive irradiation with slow neutrons is illustrated in Figure 16, in which the heavy line indicates the principal path of neutron capture and negative beta-particle decay that results in successively heavier elements and higher atomic numbers. The lighter lines show subsidiary paths that augment the major path. The major path terminates at an isotope of fermium (fermium-257), because the short half-life of the next fermium isotope (fermium-258), for radioactive decay by spontaneous fission (360 microseconds), precludes its production and the production of isotopes of elements beyond fermium by this means. The lighter lines beyond indicate predictions only.

From G. Seaborg, "Elements Beyond 100, Present Status and Future Prospects," *Annual Review of Nuclear Science,* vol. 18, p. 119 (1968)

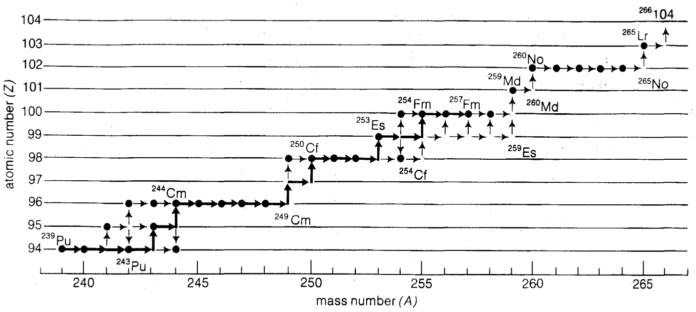


Figure 16: Nuclear reactions for the production of heavy elements by intensive slow neutron irradiation (see text).

Heavy isotopes of some transuranium elements are produced in nuclear explosions. Underground detonations of nuclear explosive devices during the late 1960s resulted in the production of significant quantities of einsteinium and fermium isotopes, which were separated from rock debris by mining techniques and chemical processing. Again, the heaviest isotope found was that of fermium-257. Why this is the case is not yet completely understood.

An important method of synthesizing transuranium isotopes, currently as well as historically, is by bombarding heavy element targets with charged particles, rather than neutrons, from accelerators. For the synthesis of elements with atomic numbers of 101 or greater, so-called heavy ions (*i.e.,* charged particles heavier than the helium nuclei, mentioned above as alpha particles) have been used.

Accelerators for superheavy elements

New accelerators, or modifications of existing ones, are necessary in order to accelerate the heavy ions needed to produce the superheavy elements. Targets and projectiles relatively rich in neutrons are required so that the resulting superheavy nuclei will have sufficiently high neutron numbers; too low a neutron number renders the nucleus extremely unstable and unobservable because of its resultingly short half-life.

NUCLEAR PROPERTIES

Isotopes of the transuranium elements are radioactive in the usual ways: they decay by emitting alpha particles, beta particles, gamma rays; and they also fission, or break up, spontaneously. In Table 28, significant nuclear properties of certain important isotopes are listed. Only the principal mode of decay is given, though in many cases other modes of decay also are exhibited by the isotope. In particular, with the isotope californium-252, alpha-particle decay is important because it determines the half-life, but the expected applications of the isotope exploit its spontaneous fission decay that produces an enormous neutron output. Other isotopes, such as plutonium-238, are useful because of their relatively large thermal power output during decay (given in Table 28 in watts per gram). Research on the chemical and solid-state properties of these elements and their compounds obviously requires that isotopes with long half-lives be used. Isotopes of plutonium and curium, for example, are particularly desirable from this point of view. The discovery of the mendelevium isotope, with a half-life of 54 days (for the emission of alpha particles) has raised hopes that all elements, at least through atomic number 101, will have some isotopes with half-lives long enough so that some of their macroscopic properties can be measured. In Table 28, the specific activities, such as disintegrations per minute per unit weight, are given for those elements that can be produced in nuclear reactors (formerly called atomic piles). Beyond element 100 the isotopes must be produced by charged-particle reactions using particle accelerators, with the result that only relatively few atoms can be made at any one time.

Nuclear structure and stability. Although the decay properties of the transuranium elements are important with regard to the potential application of the elements, these elements have been studied largely to develop a fundamental understanding of nuclear reactions and nuclear and atomic structure. Study of the known transuranium elements also helps in predicting the properties of yet undiscovered isotopes and elements as a guide to the researcher who can then design experiments to prepare and

Table 28: Nuclear Properties of Selected Transuranium Element Isotopes

name and mass	principal decay mode	half-life	specific activity	
			d/m/μg*	watts/gram
Neptunium-237	alpha	2.14×10^6 years	1.56×10^3	2.07×10^{-5}
Plutonium-238	alpha	86 years	3.86×10^7	0.570
Plutonium-239	alpha	2.44×10^4 years	1.36×10^5	$\sim 1.91 \times 10^{-3}$
Plutonium-242	alpha	3.8×10^5 years	8.63×10^3	1.13×10^{-4}
Plutonium-244	alpha	8×10^7 years	39.7	4.93×10^{-7}
Americium-241	alpha	433 years	7.59×10^6	0.114
Americium-243	alpha	7.4×10^3 years	4.24×10^5	6.45×10^{-3}
Curium-242	alpha	163 days	7.30×10^9	122
Curium-244	alpha	17.6 years	1.79×10^8	2.83
Curium-247	alpha	1.6×10^7 years	1.96×10^2	$\sim 2.8 \times 10^{-6}$
Curium-248	alpha	4.7×10^5 years	8.34×10^3	5.32×10^{-4}
Berkelium-249	beta (minus)	314 days	3.71×10^9	0.358
Californium-249	alpha	360 years	9.06×10^6	0.152
Californium-252	alpha	2.65 years	1.19×10^9	39
Einsteinium-253	alpha	20 days	5.7×10^{10}	1,000
Fermium-257	alpha	80 days	1.2×10^{10}	~200
Mendelevium-256	electron capture	1.5 hours		
Mendelevium-258	alpha	54 days		
Nobelium-259	alpha	1.5 hours		
Lawrencium-260	alpha	180 seconds		
Element 104-261	alpha	70 seconds		
Element 105-262	alpha	40 seconds		
Element 106-263	alpha	0.9 second		

*Disintegrations per minute per microgram.

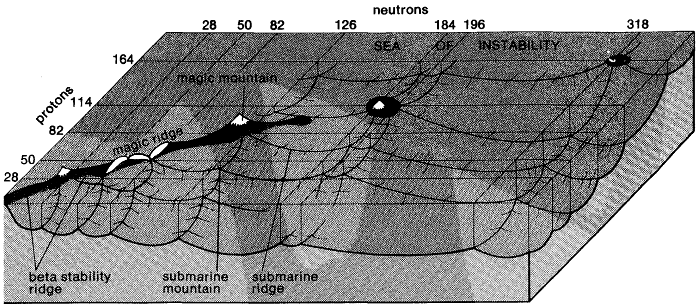


Figure 17: Known and predicted regions of nuclear stability, surrounded by a "sea" of instability.

From G. Seaborg, *Proceedings of The Robert A. Welch Foundation Conference XIII: Transuranium Elements* (November 17, 1969)

identify them. As shown in the allegorical Figure 17, the known isotopes of the elements can be represented graphically with the number of protons in the nucleus along the left-hand axis and the number of neutrons plotted on the axis along the bottom. The relative stabilities of the isotopes are indicated by their relative heights. In such representation, the known isotopes resemble a peninsula rising above a sea of instability. The most stable isotopes, appearing as mountain tops, occur at specific values called magic numbers.

Magic numbers

The magic numbers derive from calculations of the energy distribution based on the theoretical structure of the nucleus. According to theory, neutrons and protons (collectively, nucleons) are arranged within the nucleus in shells that are able to accommodate only fixed maximum numbers of them; when the shells are closed (*i.e.,* unable to accept any more nucleons) the nucleus is much more stable than when the shells are only partially filled. The number of nucleons in the closed shells yields the magic numbers. As the proton and neutron numbers depart further and further from the magic numbers, the nuclei are relatively less stable.

As the highest atomic numbers are reached, decay by alpha-particle emission and spontaneous fission sets in (see below). At some point the peninsula is terminated. With experimental methods available in the early 1980s, it appeared that the point at which the peninsula of stability disappeared lay in the vicinity of the element with atomic number 108. There has been, however, considerable speculation, based on a number of theoretical calculations, that two islands of stability might also exist. The first is expected to extend from about element 110 to element 126, while the second would be centred on element 164. With heavy-ion accelerators, experimenters might be expected to find such islands, especially the first.

Processes of nuclear decay. The correlation and prediction of nuclear properties in the transuranium region are based on systematics (that is, extensions of observed relationships) and on the development of theoretical models of nuclear structure. The development of structural theories of the nucleus has proceeded rather rapidly, in part because valid parallels with atomic and molecular theory can be drawn. At the same time, systematics has been successfully used in the case of alpha-particle decay by noting regularities in experimental alpha-particle decay energies and half-lives as a function of mass and atomic number. Similarly, for beta-particle decay, nuclear thermodynamics gives useful results. Half-lives for spontaneous fission also can be predicted by the application of systematics.

Alpha-particle decay. A nucleus can decay to an alpha particle (helium nucleus) plus a daughter product if the mass of the nucleus is greater than the sum of the mass of the daughter product and the mass of the alpha particle; *i.e.,* if some mass is lost during the transformation. The amount of matter defined by the difference between reacting mass and product mass is transformed into energy and is released with the alpha particle. The relationship is given by Einstein's equation: $E = mc^2$, in which the product of the mass (m) and the square of the velocity of light (c) equals the energy (E) produced by the transformation of that mass into energy. It can be shown that because of the inequality between the mass of a nucleus and the masses of the products, most nuclei beyond about the middle of the periodic table are likely to be unstable because of the emission of alpha particles. In practice, however, because of the reaction rate (*i.e.,* speed of reaction), decay by ejection of an alpha particle is important only with the heavier elements. Indeed, beyond bismuth (element 83) the predominant mode of decay is by alpha-particle emission, although at californium (element 98) an alternate process, called spontaneous fission, becomes possible (as a result of changes in energy balances) and begins to compete favourably with alpha-particle emission. In addition, with a number of transuranium nuclei, a third process, beta-particle decay (emission of a negative electron by the nucleus or the capture by the nucleus of an orbital electron) is of considerable importance.

Predicted half-lives

The regularities in the alpha-particle-decay process that have been noted from experimental data have been used to predict half-lives of as yet undiscovered isotopes. Such predicted half-lives are essential for experiments designed to discover new elements and new isotopes, because the experiment must take the expected half-life into account. The regularities in alpha-particle-decay energies can be plotted in a graph and, since the alpha-particle-decay half-life depends in a regular way on the alpha-particle-decay energy, the graph can be used to obtain the estimated half-lives of undiscovered elements and isotopes.

Beta-particle decay. In elements lighter than lead, beta-particle decay—emission of either negative electrons (usually called simply electrons) or positive electrons (positrons)—is the main type of decay observed. In the transuranium elements, alpha-particle decay and spontaneous fission occur, in addition to decay, by emission of negative electrons or by capture of orbital electrons. (Positron emission has not been observed in transuranium elements.) When the beta-particle decay processes are absent in transuranium isotopes, the isotopes are said to be stable to beta decay.

Decay by spontaneous fission. Regularities also are observed for spontaneous fission in the very heavy element region. If the half-life of spontaneous fission is plotted against the ratio of the square of the number of protons in the nucleus (symbolized by Z) divided by the mass of the nucleus (symbolized by A)—*i.e.,* the ratio Z^2/A—then a regular pattern results for nuclei with even numbers of both neutrons and protons (even–even nuclei). Although this uniformity allows very rough predictions of half-lives for undiscovered isotopes, the methods actually employed are considerably more sophisticated. Nuclei with odd numbers of neutrons or protons exhibit longer half-lives for spontaneous fission than those shown in the figure.

Odd–odd and even–even isotopes

The results of study of half-life systematics for alpha-particle, negative beta-particle, and spontaneous-fission decay in the near region of undiscovered transuranium elements can be plotted in graphs for even–even nuclei, and for nuclei with an odd number of protons or neutrons, and for odd–odd nuclei (those with odd numbers for both protons and neutrons). These predicted values are in the general range of experimentally determined half-lives and correctly indicate trends, but individual points may differ appreciably from known experimental data. Such graphs show that isotopes with odd numbers of neutrons or protons have longer half-lives for alpha-particle decay than do neighbouring even–even isotopes.

Nuclear structure and shape. *Liquid-drop models.* From the theoretical point of view, spontaneous-fission half-lives are generally treated by assuming that the nucleus has the properties of a liquid drop whose interior is inhomogeneous. The inhomogeneity arises from the fact that the neutrons and protons within the nucleus are more likely to be found inside rather than outside certain regions of the nucleus, called shells. This circumstance requires that a shell correction be applied to the ordinary homogeneous liquid-drop model of the nucleus. Nuclei with the exact number (or close to the exact number) of neutrons and protons dictated by closed shells (magic numbers, see above) have spherical shapes.

Since many transuranium nuclei have numbers of neutrons and protons different from the magic numbers and thus are nonspherical, considerable theoretical work has been done to describe the motions of the nucleons in their orbitals outside the spherical closed shells. These orbitals are important in explaining and predicting some of the nuclear properties of the transuranium and superheavy elements.

Two stable states for each nucleus

Nuclear-shape isomers. The mutual interaction of fission theory and experiment brought about the discovery and interpretation of fission isomers. At Dubna, U.S.S.R., in 1962, americium-242 was produced in a new form that decayed with a spontaneous-fission half-life of 14 milliseconds, or about 10^{21} times shorter than the half-life of the ordinary form of that isotope. Subsequently, more than 50 other examples of this type of behaviour were found in the transuranium region. The nature of these new forms of spontaneously fissioning nuclei was believed to be explainable, in general terms at least, by the idea that the nuclei possess greatly distorted but quasi-stable nuclear shapes. The greatly distorted shapes are called isomeric states, and these new forms of nuclear matter are consequently called shape isomers. As mentioned earlier, calculations relating to spontaneous fission involve treating the nucleus as though it were an inhomogeneous liquid drop, and in practice this is done by incorporating a correction for shells in the homogeneous liquid-drop model. The theorist found an apparently reasonable way to amalgamate the shell and liquid-drop energies. The remarkable result obtained through the use of this method reveals that nuclei in the region of thorium through curium possess two energetically stable states with two different nuclear shapes. This theoretical result furnished a most natural explanation for the new form of fission, first discovered in americium-242.

This interpretation of a new nuclear structure is of great importance, but it has significance far beyond itself because the theoretical method and other novel approaches to calculation of nuclear stability have been used to predict islands of stability beyond the point at which the peninsula of Figure 17 disappears into the sea of instability.

EXTENSION OF THE PERIODIC TABLE: THE SUPERHEAVY ELEMENTS

Possible existence of superheavy elements. According to these calculations, islands of stability will occur over the region of atomic numbers from approximately 110 to 126 and also at around 164. The so-called superheavy elements that exist at these islands of stability are predicted to have neutron numbers ranging from about 184 to 318. The nucleus most stable against all forms of decay is thought by many to be element 110, of mass 298, although the assignment of a specific atomic number and mass number is speculative. This particular nuclide (isotopic species) may have an overall half-life (although unlikely) of 10^8 years. Other hypothetical nuclides that are predicted to have exceptional stability are those with atomic numbers 114 and 124; when the neutron number of 184 is added to these numbers of protons, mass numbers of 298 and 308, respectively, are obtained. If these projections are correct, it is considered possible that superheavy elements exist in nature, and many scientific groups are looking for such elements.

The possibility of natural occurrence depends not only on the length of the half-life, but also on whether they were created in the first place. The discussion of the possible manufacture of these elements is as difficult and speculative as the discussion of their half-lives. At any rate, in the late 1960s, astronomers first observed a new kind of celestial object called a pulsar. It was thought that the most reasonable model for a pulsar was a rapidly rotating neutron star. The rapidly rotating electromagnetic and electrostatic fields associated with such a star should be able to accelerate charged particles to the velocities found in cosmic rays. If cosmic rays should originate in neutron stars, their composition might include superheavy elements. Unconfirmed evidence suggesting that elements in the atomic number range of 105 to 110 exist in cosmic rays was reported in 1968.

The existence of superheavy elements in meteorites or in terrestrial minerals depends on the probability of their having been formed by the same fundamental processes that created the other elements. Consequently, it is worthwhile to look for superheavy elements in meteorites and terrestrial minerals, since their presence or absence may tell something about the origin of the universe. Since the early 1970s, searches have uncovered certain anomalies that might be attributed to superheavy elements, but the results cannot be said to furnish proof of their existence.

Electronic structure of atoms

Superheavy elements and their predicted properties. The postulated nuclear islands of stability are important to chemistry. The periodic table of the elements classifies a wealth of physical and chemical properties, and study of the chemical properties of the superheavy elements would show how far the classification scheme of the table could be extended on the basis of nuclear islands of stability. Such studies would shed new light on the underlying properties of electrons orbiting the nucleus because it is these properties that produce the periodic system. The positions of superheavy elements in the periodic table ultimately would be determined by the characteristic energies of the electrons of their atoms, especially the valence electrons. Complex calculations have predicted meaningful distribution of electrons in orbitals for a number of superheavy elements. Results for elements 104–121 are given in Table 29, the configurations being those that the atoms have when they are at their lowest energy level, called the ground state.

Table 29: Calculated Electronic Ground States for Some Superheavy Elements

atomic number	electronic structure of two outermost shells*	atomic number	electronic structure of outermost shells†
104	$6d^27s^2$	113	$7s^27p^1$
105	$6d^37s^2$	114	$7s^27p^2$
106	$6d^47s^2$	115	$7s^27p^3$
107	$6d^57s^2$	116	$7s^27p^4$
108	$6d^67s^2$	117	$7s^27p^5$
109	$6d^77s^2$	118	$7s^27p^6$
110	$6d^87s^2$	119	$7s^27p^68s^1$
111	$6d^97s^2$	120	$7s^27p^68s^2$
112	$6d^{10}7s^2$	121	$7s^27p^67d^18s^2$

*Added to the structure of radon, element 86, plus orbitals $5f^{14}$.
†Added to the structure of radon, element 86, plus orbitals $5f^{14}6d^{10}$.

Elements 113 and 114. These calculations of electronic structure permit predictions of detailed physical and chemical properties of some superheavy elements. If, for example, the structure of the periodic system (Figure 18) remains predictable to higher atomic numbers, then element 113 will be in the same group of elements as boron, aluminum, gallium, indium, and thallium; and element 114 will be in the group with carbon, silicon, germanium, tin, and lead. Element 113 usually is called ekathallium, and element 114 is called ekalead. If they ever are discovered, of course, these elements will be given proper names of their own. Computer calculations of the character and energy levels of possible valence electrons in the atoms of these two superheavy elements have substantiated the placement of the elements in their expected positions. Extrapolations of properties from elements with

Table 30: Some Predicted Properties of Elements 113 and 114

	element 113 (ekathallium)	element 114 (ekalead)
Chemical group	III	IV
Atomic weight	297	298
Most stable oxidation state	+1	+2
Oxidation potential, V	−0.6 $M \rightarrow M^+ + e^-$	−0.8 $M \rightarrow M^{2+} + 2e^-$
Metallic radius, Å	1.75	1.85
Ionic radius, Å	1.48	1.31
First ionization potential, eV	7.4	8.5
Second ionization potential, eV	. . .	16.8
Density, g/cm³	16	14
Atomic volume, cm²/mole	18	21
Boiling point, °C	1,100	150
Melting point, °C	430	70
Heat of sublimation, kcal/mole	34	10
Heat of vaporization, kcal/mole	31	9
Debye temperature, °K	70	46
Entropy, eu/mole (25°C)	17	20

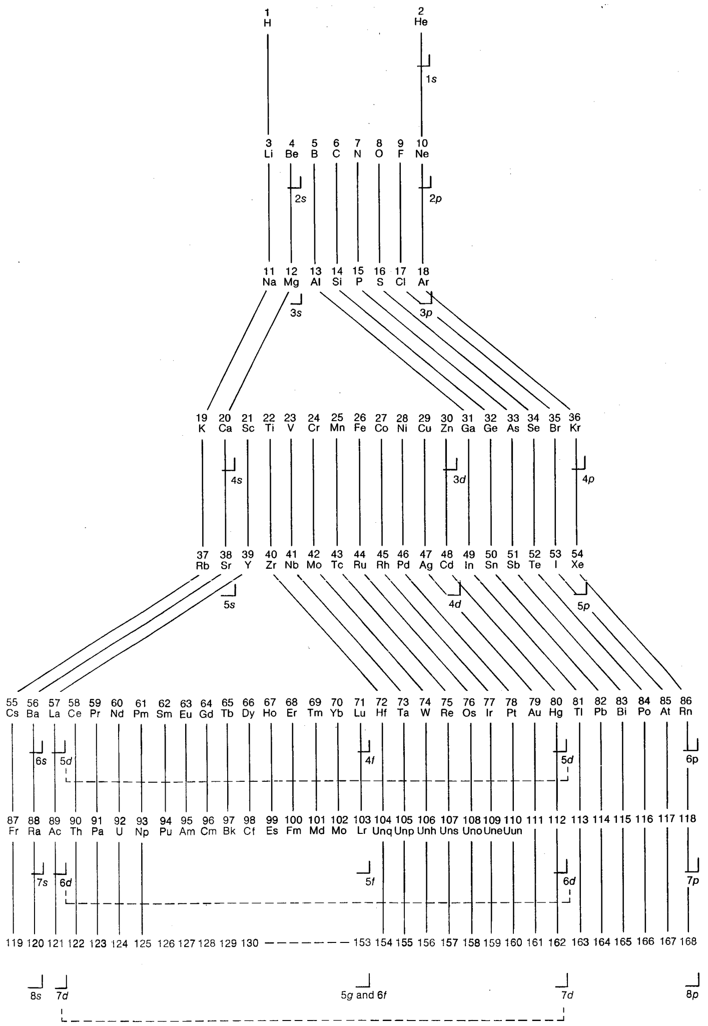


Figure 18: Modified form of a periodic table showing known and predicted electron shells.

From G. Seaborg, *Proceedings of The Robert A. Welch Foundation Conference XIII: Transuranium Elements* (November 17, 1969)

lower numbers to elements 113 and 114 can then be made within the usual limitations of the periodic table. In Table 30, the results of such extrapolations are given. Although, in many cases, theoretical calculations are combined with extrapolation, the fundamental method involved is to plot the value of a given property of each member of the group against the appropriate row of the periodic table. The property is then extrapolated to the seventh row, the row containing elements 113 and 114. The method is illustrated in Figure 19 for estimating the melting point of element 113.

Bonding properties

As explained above, the bonding property of an element can be expressed by the energy required to shift a bonding, or valence, electron. This energy can be expressed in various ways, one of which is a relative value called the oxidation potential. The relative stabilities of possible oxidation states of an element represent what is probably that element's most important chemical property. The oxidation number of the atom of an element indicates the number of its orbiting electrons available for chemical bonds or actually involved in bonds with other atoms, as in a molecule or in a crystal. When an atom is capable of several kinds of bonding arrangement, using a different number of electrons for each kind, it has that many oxidation numbers, or oxidation states. The prediction of oxidation states can be illustrated with element 114, ekalead, which occurs in Group IVa of the periodic table. The outstanding periodic characteristic of the Group IVa elements is their tendency to go from a +4, or tetrapositive, oxidation state to a +2, or dipositive, state as the atomic number increases. Thus, carbon and silicon are very stable in the tetrapositive state, whereas germanium shows a weak dipositive state in addition to its strong tetrapositive state. Tin chemistry is characterized by about equal stability in the tetrapositive and dipositive states, and lead chemistry is dominated by the dipositive state—the element showing only weak tetrapositive properties. Extrapolation in the periodic table to the seventh row, then, results in a predicted most stable dipositive oxidation state for element 114. This result is supported by valence bond theory and by extrapolations of thermodynamic data.

Other superheavy elements. Less detailed predictions have been made for other superheavy elements. Element 117, for example, is expected to be a member of the halogen series, which is the group composed of fluorine, chlorine, bromine, iodine, and astatine. Solid element 117 should be metallic in appearance, as is astatine, but it is expected that, instead of the minus one (−1) oxidation-state characteristic of the natural halogens, it will show +3, +5, and +7 oxidation states. It should also form stable interhalogen compounds with fluorine, chlorine, and bromine.

Computer calculations of element 118

Computer calculations suggest that element 118 should have the closed-shell electronic configuration of the noble gas elements helium, neon, argon, krypton, xenon, and radon. The element should be the most electropositive of the noble gases, and, therefore, the existence of a (partially ionized) difluoride of the element 118 is predicted. A tetrafluoride and an oxide of the type formed by xenon (XeO_4) are also expected.

Although detailed predictions of the chemistry of element 119 have not as yet been completed, it is expected to be a typical alkali metal with a plus one (+1) oxidation state. The energetic properties of its valence electron, the 8*s* electron, suggest that its first ionization potential will be higher than the oxidation potential predicted by simple extrapolation, so that the element may be more like potassium than cesium in its chemistry. This higher energy will cause the metallic and ionic radii to be smaller than simple extrapolation would indicate.

Element 120 would be expected to be a typical alkaline-earth element. As with element 119, the ionization potentials should be raised to higher levels than the normal family trend would indicate and should also make the metallic and ionic radii smaller. These changes should make the chemistry of element 120 similar to that of calcium and strontium.

Element 121 should be similar in its chemical properties

From O. Keller Jr. et al., *Journal of Physical Chemistry*, vol. 74 (1970); American Chemical Society

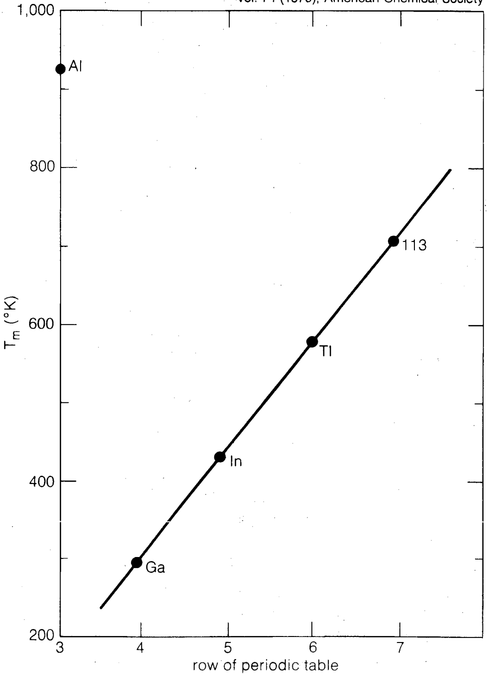


Figure 19: Melting points (Tm) in degrees kelvin (K) of Group III elements extrapolated to element 113 (see text).

to lanthanum and actinium, but detailed properties have not been predicted.

Superactinide series. It is probable, in a formal sense at least, that element 122 will begin another series of elements in which each successive electron is added to a deep inner orbital, in a manner similar (see Figure 18) to that found in the lanthanide and actinide series (see above *Rare-earth elements* and *Actinide elements*). Such a series, which would be listed under the actinide series in the periodic table, should consist of 32 elements, ending in the neighbourhood of element 153 and resulting primarily from the filling of the 5*g* and 6*f* inner electron shells.

Not every element of this new series would correspond to an actinide (or lanthanide) element on a 1:1 basis, and prediction of the chemistry of the members of the series is a very complex problem. The difficulty arises partly because of uncertainty of the exact point at which the energetically similar 5*g* and 6*f* orbitals begin to fill, and partly because calculations indicate that the 8*p* and 7*d* orbitals may be very close in energy to the 5*g* and 6*f* orbitals. (The number 8 shell is outermost from the nucleus and in each shell the *s* orbital is the innermost.) These orbitals may all be filled, then, in a commingling fashion, resulting in a series of elements that show multiple, barely distinguishable oxidation states. The electronic basis for the periodicity shown in Figure 18 will then no longer be present.

As shown, element 153 will be the last member of the superactinide series, at least in a formal sense. The prediction of properties on the basis of an orderly extrapolation appears to be of doubtful validity, however, in this heavy-element region of the periodic table. In still higher elements, the closely spaced energy levels are expected to make multiple oxidation states the rule. The placement of the elements in the heaviest portion of the periodic table as shown in Figure 18 is, therefore, probably also of only formal significance.

Maximum atomic number

End of the periodic table. At some point the stability of the orbital electrons in the ordinary sense must be destroyed as more and more protons are added to the nucleus. There is, therefore, in the ordinary chemical sense, a critical atomic number, or range of atomic numbers, which represents the end of the periodic table. This end, it should be noted, is separate, at least philosophically, from the question of stability of the nucleus itself; *i.e.,* nuclear stability is not the same as stability of the electron shells. The maximum atomic number, according to current theories, lies somewhere between elements 170 and 210, as indicated above.

CHARACTERIZATION AND IDENTIFICATION

Two important factors provided the key to the discovery and identification of many of the transuranium elements: one was knowledge of the lanthanide series coupled with the concept that it could be paralleled in the actinide series, and the other was the technique of separating elements with similar properties from a mixture by using the principle of ion exchange (see CHEMICAL REACTIONS). Ion-exchange reactions depend on the fact that some complex molecules have a charge that will attract ions of the opposite charge, hold them, and then exchange them for other ions of the same charge when brought in contact with them. The actinide concept, proposed in 1944, stated that the transuranium elements were part of a series of elements that started at thorium and that the chemistry of the series would be similar to that of an earlier transition-element series, the lanthanides. By the use of ion-exchange reactions that are highly specific, many of the transuranium elements have been separated and identified. The tripositive ions of the lanthanides and actinides are eluted (washed) from a cation-exchange resin (a resin that has a negative charge and will hold positive ions or cations) using a particular organic ion, alpha-hydroxyisobutyrate. The striking similarity between elution patterns of the two groups constitutes strong support for the actinide concept. A certain measure of stability of the dipositive state for nobelium was predicted by the actinide concept. A position was predicted for lawrencium, but for this element the yields have been so small and the half-lives so short that its positive identification by ion exchange has not been achieved. The tripositive state of lawrencium has been confirmed by a very rapid solvent-exchange experiment in which the lawrencium displayed the behaviour of the tripositive actinides and not that of the dipositive nobelium or radium, again in accord with the predictions of the actinide concept.

Non-chemical identification and separation

When the yields of a new element are small and its half-life is short, chemical identification and characterization are frequently not possible. In such cases the atomic number is deduced from the method of production through nuclear reactions, from the parent-daughter relationship of the new element to known elements of lower atomic number resulting from its nuclear decay, and from its nuclear-decay systematics that cannot be attributed to any known nuclides. Additionally, the variation in the yield of the new element is noted when the bombarding energy is changed or when the target or projectile or both are changed.

Separation of the product nuclide from the target has been accomplished in the discoveries of elements 101 through 106 by a recoil collection method. When the target nucleus is struck by a heavy ion projectile, the product nucleus recoils out of the very thin target and is either attracted to a substrate by an electrostatic potential or is swept onto a substrate by a jet of helium gas. The new element is then in a position to be observed and characterized by suitable detection techniques, essentially free of the parent isotope.

It is desirable, though not essential, that the mass number of the new element be established by evidence related to its mode of production, or to its parent-daughter relationship through radioactive decay to a radioactive isotope of known mass number. When weighable quantities of a transuranium element are available, more extensive characterization experiments can be performed. The most important of these experiments is the preparation of the metal, frequently done by high-temperature reduction of the fluoride of the transuranium element with an alkali or alkaline-earth metal. Another method used for preparation of larger (gram) quantities of high purity is electrolytic reduction of the chloride of the transuranium element. Physical characterization of these metal samples includes determination of the density, melting point, vapour pressure, boiling point, hardness, and other properties. X-ray diffraction measurements permit the determination of the crystal structure and calculation of the metallic radius and metallic valence. Chemical characterization includes a determination of the reactivity of the metal with other substances and the chemical stability of the compounds formed. Also of importance are the oxidation states and chemical-bonding properties of the elements in its compounds.

PRACTICAL APPLICATIONS OF TRANSURANIUM ISOTOPES

Three transuranium isotopes, of plutonium, americium, and californium (plutonium-238, americium-241, and californium-252), have demonstrated substantial practical applications. One gram of plutonium-238 produces approximately 0.56 watt of thermal power, primarily from alpha-particle decay, and this property has been used in space exploration to provide energy for small thermoelectric-power units (see above *Actinide elements*).

Americium-241 has a predominant gamma-ray energy (60 keV) and a long half-life (433 years) for decay by the emission of alpha particles, which make it particularly useful for a wide range of industrial gauging applications and the diagnosis of thyroid disorders. When mixed with beryllium it generates neutrons at the rate of 1.0×10^7 neutrons per second per gram of americium-241. The mixture is designated ^{241}Am–Be and a large number of such sources are in worldwide daily use in oil-well logging operations; *i.e.,* to find how much oil a well is producing in a given time span, as a day. Californium-252 is an intense neutron source: one gram emits 2.4×10^{12} neutrons per second. This isotope has been tested for applications in neutron activation analysis, neutron radiography, and portable sources for field use in mineral prospecting and oil-well logging. (G.T.S./Ed.)

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(G.T.S.)

Chemical Reactions

The phenomenon of change, noted by everyone almost continuously and said by some to be the only unchanging aspect of the material universe, is the context in which all of the sciences find their place, including the study of chemical reactions. A chemical reaction is a process of change during which either two or more substances both change or one substance changes into at least two other substances. Thus, when wood burns, the substances present initially, wood and oxygen in the atmosphere, are converted in a chemical reaction to water vapour, carbon dioxide, and ash. All combustions are chemical reactions. Other types of familiar chemical reactions include decay, fermentation, the hardening of cement, the development of a latent image in an exposed photographic film, the tarnishing of silver, the corrosion of steel, the evolution of gas when vinegar and soda are mixed, the synthesis of nylon, and the digestion of food.

In a general sense, material substances can undergo change in three ways: a change of position, called movement; a change of form, such as the freezing of liquid water; and a change of substance, a chemical reaction. Some classify changes of form as chemical reactions, but, historically, the term chemical reaction has been applied only to changes of substance. The application to change of form is discussed below. Using the historical definition, each different chemical reaction displays the same unique characteristics.

Transformation. Thus, in a chemical reaction the substances originally present disappear, and substances that were not initially present appear. Factually, however, it is more descriptive to say that properties that were initially observable are no longer observed, and properties not originally observable are now noted. In combustion, a substance called wood, with its unique properties (fibrous, less dense than water, generally light- or dark-brown in colour), and oxygen, with its unique properties, all of which were capable of detection before combustion occurred, cannot be detected after combustion. Instead, after combustion, only the properties of water vapour, carbon dioxide, and ash can be detected.

Conservation of mass. Chemical reactions display another essential quality. Although substances change in a chemical reaction, within limits that can be measured the total mass does not change. That is, the mass of wood and oxygen that disappears in a combustion is equal to the mass of water vapour, carbon dioxide, smoke, and ash that appears. In ordinary chemical reactions, this loss of mass as the reactants vanish is equal to the gain of mass as the new substances form. Because this effect is universal, it has been presumed to be fundamental, an indication that there is some common, universally applicable reason that applies to all material substances.

Fixed composition. Further, substances formed in chemical reactions display another universal but more specific characteristic. For example, water is formed as one of the products in hundreds of quite different chemical reactions, but, without exception, it contains only the elements hydrogen and oxygen and always in the same proportion: the mass of oxygen is eight times as great as the mass of hydrogen. That is, the composition of water is independent of the particular chemical reaction by which it might have been formed. Parallel phenomena are observed with other substances: whenever carbon dioxide is formed as a product, the mass ratio of carbon to oxygen is always 3 to 8; for ammonia (a gas, not the aqueous solution called household), the ratio of its constituent elements, nitrogen and hydrogen, is always 14 to 3. (Precise measurements reveal minor variations in the mass ratio that are understood in terms related not to chemical reaction but to nuclear composition; see ATOMS: *Isotopes.*)

Energy effects. Finally, in all but a very few chemical reactions, energy is either absorbed or evolved. It is often evolved as heat, as in combustion, but also in other forms, such as electrical energy during the chemical reaction in a battery when it is switched on to operate a flashlight or portable radio. The amount of energy evolved depends on the mass of the products formed. For twice the mass of product formed, in two otherwise identical chemical reactions, exactly twice as much energy is evolved. The same relationship applies to reactions during which energy is absorbed rather than evolved.

A chemical reaction can now be defined more explicitly as a process of change in which the substances originally present, called reagents, are changed into substances with other properties, called products, in such a way that, first, there is no observable change in the total mass; second, whenever the same product is formed by a different process of change, that product exhibits the same mass ratio of components; and, third, almost always, energy is either absorbed or evolved in an amount that is directly related to the mass of the products formed.

These facts have two important consequences. First, they can be used to infer universally applicable, theoretical principles to account for the reasons that chemical reactions occur and for some of the details of the process, thereby increasing understanding of the material universe. Second, these facts and the related uncertain but usable theories can be used to alter man's environment for either his benefit or his detriment. Examples abound. The chemical industry manufactures beneficial substances, such as polymeric fibres (*e.g.,* nylon) and elastomers (synthetic rubber), dyes, other polymers (plastics), metallic alloys, fertilizers, paints, insecticides, drugs, ceramics, and fuels; but these manufactures also cause undesired effects upon man's ecological system, creating problems that demand further application of chemistry as well as other branches of knowledge.

This article is divided into the following sections:

Introductory survey

GROWTH OF MAJOR THEORIES CONCERNING CHEMICAL REACTIONS

Phlogiston theory

Chemical reactions were known to prehistoric man, and it is possible to conjecture that he speculated upon their cause. In ancient Greek civilization this speculation led to a qualitative supposition that matter is perhaps composed of individual particles. Not until late in the 17th century, however, did theory and philosophical speculation become conjoined in the first of many far-reaching attempts to explain chemical reactions. The German chemist and physician Georg Ernst Stahl, basing his ideas upon earlier and less specific suggestions, postulated that during combustion a substance, which he named phlogiston, escaped from the burning fuel into the air. Thus if a drinking glass is inverted over a burning candle, the flame dies out in a short time, because, it was postulated, the air within the glass becomes saturated with phlogiston. According to these ideas, because ash remains after combustion, wood is composed of phlogiston and ash. Because no ash is left when a candle burns, candle wax is obviously pure phlogiston. Therefore it should be possible to prepare wood in a chemical reaction between wood ash and candle wax. This prediction of the phlogiston theory failed. It failed in more startling ways, as well. When metals burn, for example, their ash is greater in mass than the original metal. To explain this, it was suggested that phlogiston possessed the quality of negative mass, which is a direct contradiction of its other postulated properties.

Additional evidence that led to the downfall of the phlogiston theory was accumulated by several workers, among whom Carl Wilhelm Scheele, in Sweden, and Joseph Priestley, first in England and later in North America, can be mentioned particularly. Independently, each discovered a gas (actually oxygen), but both failed to recognize the significance of their discovery. Priestley in particular maintained his belief in phlogiston until his death. At about the same time, however, near the end of the third quarter of the 18th century, Antoine-Laurent Lavoisier, in France, independently discovered the same gas, and he recognized its significance. He postulated that combustion was a chemical reaction involving two substances, one, a component of the air, which he named *oxigine* (oxygen), and the other a combustible substance, such as candle wax, wood, or metal.

Dalton's atomic theory

Shortly after the beginning of the 19th century, John Dalton in England postulated that matter is composed of small, indivisible, and unalterable particles called atoms and that in a typical chemical reaction groups of atoms initially conjoined in some way became disassociated and then rejoined in new arrangements. The observed disappearance of properties and the appearance of new properties were thus explained by the theory that the properties of a substance depend upon its atomic composition. Centuries earlier, others had expressed ideas that were very similar, but Dalton was able to convince his contemporaries of the probable validity of his postulate.

Of equal importance was Dalton's concept that each atom of any single element was identical in every respect to every other atom of that same element. As a consequence, the masses of the atoms are quantitatively revealed in the ratio of the masses of the components of a substance. Thus, because the mass of oxygen in water is eight times that of the mass of hydrogen, it can be said that one atom of oxygen is eight times as massive as one atom of hydrogen, on condition that one particle (today denoted by the word molecule) of water is composed of one atom each of hydrogen and oxygen. As Dalton recognized, however, one particle of water might be composed of some other number of hydrogen and oxygen atoms—for example, two atoms of hydrogen and one atom of oxygen—then the relative masses of oxygen and hydrogen atoms would be 8 to $^1/_2$ or, in whole numbers, 16 to 1.

During this same period, other facts obtained in laboratory studies, particularly those of Joseph-Louis Gay-Lussac, in France, suggested that, although the masses of gases that reacted with each other showed no simple consistency, the volumes of those reacting gases did demonstrate a certain simplicity. For example, eight grams of oxygen reacted with one gram of hydrogen, and 14 grams of nitrogen reacted with three grams of hydrogen; but, when measured at equal temperatures and pressures, the volume of hydrogen was found to be exactly twice that of the volume of oxygen in the reaction to form water, and the volume of hydrogen was found to be exactly three times that of nitrogen in the reaction to form the gas ammonia. Apparently, this small integer volume relationship hinted at an as yet unknown fundamental property of matter. Amedeo Avogadro in Italy suggested that this property could be explained by the assertion: All gases at the same temperature, pressure, and volume will contain the same number of particles, which may or may not be single atoms. His suggestion was not well received, but Avogadro's pupil, Stanislao Cannizzaro, carried out further studies, in the light of which it became reasonable to state that water molecules contain two atoms of hydrogen, ammonia molecules three atoms of hydrogen, hydrogen chloride molecules one atom of hydrogen, and, for each, some number of atoms of other elements. These compositions are symbolized today by formulas in which the symbols for the elements represent single atoms. Thus, water is H_2O, the subscript 2 indicating two atoms of hydrogen and a subscript of 1 being understood for one atom of oxygen. For ammonia, the formula is NH_3: one atom of nitrogen and three atoms of hydrogen in each molecule. For hydrogen chloride (Cl being the symbol for chlorine), HCl: one atom of each. Approximately 50 years after the initial publication of Avogadro's ideas and largely through the efforts of Cannizzaro, Avogadro's hypothesis, or theory, as it is more commonly called today, was accepted by most scientists. (Modern methods of analysis have established that the actual number of molecules in a quantity of any substance equal in mass in grams to the molecular weight of the substance—*e.g.,* 18 grams of water—is 6.02×10^{23}, called Avogadro's number.)

Avogadro's number

Facts revealed by intensive laboratory examination of other properties of matter (*e.g.,* magnetism, electricity, radiations) led to theoretical conclusions that atoms are not indivisible but are themselves composed of still smaller parts, or particles. One of these parts is a negatively charged particle named the electron (the meaning of "negative charge" is still unknown except for broad generalities), and another is a positively charged particle called the nucleus. The interaction of electrons from different atoms, often conjointly as electron pairs, with the positively charged nuclei of each of the atoms, constrains them to group together in fixed ways as molecules. Not all substances, however, are molecular. For example, salt, a compound of sodium and chlorine, consists of an aggregation of charged particles called ions. In particular, salt is an aggregation of any number of sodium ions and of an equal number of chloride ions. A sodium ion is derived from a sodium atom by the loss of one electron; a chloride ion is derived from a chlorine atom by the gain of one electron. Still other nonmolecular aggregations are believed possible. Diamond is an aggregation of any large number of carbon atoms arranged in a unique three-dimensional configuration in a repetitive manner, much as the design of wallpaper is repetitive in two dimensions. In this respect, salt is like diamond; its ions are arranged in any single crystal of salt as a repetitive pattern not unlike a three-dimensional chessboard might be imagined, in which black and white squares represent sodium and chloride ions, respectively.

With this information, a more recently developed definition of a chemical reaction can be stated. Thus, if atoms are held together, or bonded, in molecules or in other types of aggregations by electron interactions, then in a chemical reaction these bonds are broken (by the absorption of energy) and, either simultaneously or subsequently, new electronic interactions develop as other bonds are formed (with the release of energy). Hence, distinct from the historically developed ideas, a chemical reaction can now be defined as a process of change in which some bonds are broken and different bonds are formed. This definition includes all processes that involve a change of substance as well as many of the processes that involve a change of form, such as the freezing of water.

ENERGY CONSIDERATIONS

If the energy absorbed in bond rupture exceeds the energy released by the formation of new bonds, then overall the chemical reaction is observed to be energy absorbing. The converse is true for cases in which the energy absorbed is less than the energy released. In a very few instances the overall absorbed and released energies are equal in magnitude.

Types of reaction. Chemical reactions can be classified into three types: exoergic (or exothermic) if, overall, energy is evolved; endoergic (or endothermic) in the converse cases; and aergic (or athermic; *i.e.,* without energy change) for rarer cases.

In every case, however, energy must be supplied to the reactants in order to initiate the breaking of bonds before other bonds can be formed because a stable bond will not of itself degenerate. In general, therefore, all chemical reactions, even exoergic or aergic, require the introduction of energy in some form from an external source in order to begin. The initiating energy, called activation energy, is sometimes supplied as heat from another, already initiated exoergic chemical reaction. Thus, to set fire to paper the activation energy can be supplied by a burning match, for which the activation energy was generated by a chemical reaction of the materials in the match head, for which activation energy was supplied as heat generated from frictional effects when the match head was rubbed upon a suitable surface.

Use of electrical energy

The required energy may be furnished, instead, as electrical energy, as for the endoergic decomposition of water into hydrogen and oxygen, the electrical energy being obtained from generators driven by turbines, which are in turn powered by falling water. In still other instances the requisite energy is supplied in the form of light, which can be thought of as consisting of discrete particles of electromagnetic energy, called photons. Photons that give rise to the sensation of red have less energy than photons that give rise to the sensation of orange, and these have less energy than photons that produce the sensation of yellow, and so on. Photons of lesser energy and photons of greater energy than those that comprise the range of visible colours are known and are called infrared and ultraviolet photons, respectively. If a photon of the particular quantity of energy needed to break a particular chemical bond passes near enough to that bond, it is probable that that photon will cease to exist and its energy will be absorbed by that bond as it breaks. Reactions that can be initiated in this manner are called, appropriately, photochemical reactions. The best known photochemical reaction is actually a series of consecutive reactions that takes place in the green leaves of plants through the influence of sunlight. These reactions, called photosynthesis, involve the consumption of carbon dioxide from the atmosphere and of water present in the plant, with the production of oxygen, released to the atmosphere, and of cellulose and starch (or sugar), which remain within the plant structure.

Catalytic reactions. For reasons not yet well understood, many chemical reactions can be initiated with a lesser activation energy than normally required when they are conducted in the presence of special foreign substances called catalysts; such a reaction is said to be catalyzed. Chlorophyll is a catalyst in the photochemical reaction of plants. Enzymes are involved as catalysts in the metabolic processes (chemical reactions) that occur in living tissue; the enzyme pepsin, present in the stomach, catalyzes the breakup of large protein molecules into smaller molecules.

A sugar cube will sputter but not burn when a match flame is applied to its surface, but very small amounts of substances called rare-earth oxides act as catalysts for this particular combustion; the oxides are present in trace amounts in tobacco ash, and a cube lightly coated with tobacco ash will burn if heated with the flame of a match.

At one time it was thought that the driving force, the cause of the spontaneity of chemical reactions, the reason, so to speak, that wood burns or cement hardens or an egg congeals when it is cooked, could be attributed to energy relationships such as those discussed above.

Entropy. It is now known that these energy relationships are indeed related to the rapidity or slowness of any particular chemical reaction, but the reason a chemical reaction occurs, at any speed, is attributed to changes in what is called the entropy both of the substances involved in the reaction and of the surroundings not otherwise involved in the chemical reaction itself. Entropy is the measure of that energy that is associated with disorder in any system; it is a concept developed in thermodynamics to take into consideration the fact that not all types of energy can be manipulated to do work. Thus, in any isolated system, entropy tends to increase; *i.e.,* the portion of energy that is not available for work is transformed into the energy of disorder. Consider, for example, the reaction in which wood combines with oxygen: water vapour, carbon dioxide, and ash are formed, and heat is evolved. The entropy of the wood and oxygen is relatively small, while the entropy of the products is larger; that is, for the substances themselves, entropy has increased. The evolved heat also causes an increase in the entropy of the surroundings that absorb that heat. Hence, overall, for the substances and for the surroundings, entropy has increased. Unless overall there is an increase in entropy, a chemical reaction cannot occur. In the synthesis of water from hydrogen and oxygen, to be described more fully below, the entropy of the product, water, is less than the entropy of the reagents, hydrogen and oxygen. The heat evolved in this reaction, however, is sufficient to increase the entropy of the surroundings more than the decrease of entropy suffered by the substances themselves; overall, that is, entropy increases. For this reason the reaction is spontaneous.

MODERN VIEWS AND CLASSIFICATION SYSTEMS

When the science of chemistry began, only the masses of the reagents and products could be measured. This early emphasis upon mass and, practically speaking, the weighing of samples influenced the theories of chemical reaction. As understanding developed, measuring instruments more sophisticated than balances and volumetric glassware were designed, particularly as a consequence of developments in physics. Today, it is possible to measure with precision and reliability a variety of subtle effects, ranging from the absorption and emission of energy as heat, as photons, or as electrical energy to detection of almost unimaginably small amounts of reagents. In general, modern chemical theoretical research is concerned with either the mechanism of a chemical reaction (called kinetics, or kinetics and mechanisms, or the dynamics of chemical reactions) or with the construction of mental or physical models of the structure of matter (called structural chemistry). In addition, applications of both new and old theory are made with the consequent introduction of new or improved substances useful in commerce. There is perhaps no single case today for which all of the theoretical details can be completely described, and only a vague distinction can be made between theoretical and applied chemistry.

Concerns of modern theoretical research

Various other classifications and types of chemical reactions exist, derived largely from a theoretical viewpoint. They are not, however, all-inclusive, as are the classifications of chemical reactions as exoergic, endoergic, and aergic. Thus, a synthesis reaction may also be an oxidation–reduction reaction. Some acid–base reactions may also be ionic reactions, or precipitation reactions.

Synthesis. A synthesis reaction, in the simplest sense, involves elements as reagents and the formation of a compound (a substance composed of more than one element) as the product, often as the only product. Iron, symbolized as Fe, reacts with sulfur, S, to form iron sulfide, FeS, as shown by this chemical equation: $Fe + S \rightarrow FeS$. Thus, the plus sign on the left symbolizes "reacts with"; the arrow signifies "forms," "produces," or "yields."

In addition to symbolizing the substances, iron, sulfur, and iron sulfide, the symbols used in a chemical equation also specify the amounts of substances that react and are produced. Thus, Fe represents 55.85 grams of iron, S represents 32.06 grams of sulfur, and FeS represents 87.91 grams of iron sulfide. That is, the chemical equation given above summarizes laboratory-measured fact: 55.85 grams of iron will react with exactly 32.06 grams of sulfur to form exactly 87.91 grams of iron sulfide. If other amounts are used, say one-fifth as much iron, 11.17 grams, no matter how much sulfur was present in excess, only one-fifth of 32.06 grams would be consumed and one-fifth as much product formed. The converse is true (in round numbers): if half as much sulfur is available, then half as much iron would react, and about 44 grams of iron sulfide would be formed. The calculation of the amounts of reagents consumed and of products formed constitutes the branch of chemistry called stoichiometry.

As a second example of synthesis, the following equation describes the synthesis of water from its elements: $2H_2 + O_2 \rightarrow 2H_2O$. Here, a new feature of a chemical equation appears, the stoichiometric factor, 2, preceding the formula for hydrogen molecules, H_2 (two atoms in each molecule), and water molecules. That is, in the laboratory it has been observed that four grams of hydrogen react with 32 grams of oxygen to produce 36 grams of water. The symbol for one gram of hydrogen in the form of atoms is H; the expression for two grams of hydrogen in molecular form is H_2. In the reaction described by the equation, then, the expression $2H_2$ represents four grams of hydrogen in the form of molecules. Analogously, the formula H_2O represents 18 grams of water molecules.

The mole unit

These quantities—H, or one gram of hydrogen atoms; H_2, or two grams of hydrogen molecules; H_2O, or 18 grams of water molecules; Fe, or 56 (approximately) grams of iron atoms; S, or 32 grams of sulfur atoms—are called moles of these substances. Thus, one mole of hydrogen atoms weighs one gram; one mole of water molecules weighs 18 grams; one mole of iron sulfide molecules weighs 87.91 grams. In each of these instances and in all others that could also have been mentioned as additional examples, the same number of particles is understood to be specified. Thus, one mole of hydrogen molecules contains 6.02×10^{23} (602,000,000,000,000,000,000,000) molecules (Avogadro's number, see above); one mole of iron atoms is 6.02×10^{23} atoms; one mole of water molecules is 6.02×10^{23} molecules; and so on. With this kind of stoichiometric emphasis, the equation describing the synthesis of water can be read as: two moles of hydrogen molecules react with one mole of oxygen molecules to form two moles of water molecules. The same stoichiometric emphasis is usually applied to the equations for other types of chemical reactions, described below.

Decomposition. Decomposition reactions are chemical reactions in which chemical species break up into simpler parts. The decomposition of the gas ammonia is represented by the equation $2NH_3 \rightarrow N_2 + 3H_2$; or, in terms of a single mole of ammonia, $NH_3 \rightarrow {}^1\!/_2N_2 + {}^3\!/_2H_2$, read as: one mole of ammonia molecules decomposes to form one-half mole of nitrogen molecules and three-halves of a mole of hydrogen molecules.

Compounds need not break down into elements in a decomposition reaction. For example, ammonium carbonate, $(NH_4)_2CO_3$, decomposes into ammonia, carbon dioxide (CO_2), and water, according to the equation $(NH_4)_2CO_3(s) \rightarrow 2NH_3(g) + CO_2(g) + H_2O(g)$.

The letters in parentheses indicate the physical states of the substances, s for solid, g for gaseous, l for liquid.

Polymerization. Polymerization reactions are not unlike synthesis reactions in that simpler substances combine to form more complex substances. The term polymerization, however, is restricted to chemical reactions in which the product is composed of many, hundreds or thousands, of the simpler reagent species. The polymerization of terephthalic acid, $HO_2C(C_6H_4)CO_2H$, with ethylene glycol, $HOCH_2CH_2OH$, to form the polymer called Dacron in fibre form or Mylar in sheet form, is represented by the equation:

$$n\,HO_2C(C_6H_4)CO_2H + n\,HOCH_2CH_2OH \rightarrow [\ldots OC(C_6H_4)CO_2CH_2CH_2O\ldots]_n + 2n\,H_2O$$

in which n signifies a large number of moles (and $2n$ twice that number of moles): the dotted extensions at either end of the repetitious polymeric molecule symbol signify further extensions of the same pattern. The polymer ends eventually with an $HO_2C(C\ldots)$ at the left and a $\ldots CH_2OH$ at the right end.

Chain reactions. A chain reaction is a series of reactions in which the product of each step is a reagent for the next. Many polymerization reactions are chain reactions. A simpler example, however, is found in the synthesis of hydrogen bromide. The overall synthesis equation is $H_2 + Br_2 \rightarrow 2HBr$. The details by which this synthesis occurs are believed to involve a series of reactions beginning with $Br_2 \rightarrow 2Br$, which is endoergic. Some of the bromine atoms will recombine, however, in the reverse exoergic reaction, $Br + Br \rightarrow Br_2$, but not all do so. If a bromine atom instead moves in such a way as to meet and interact with a hydrogen molecule, another reaction will occur: $Br + H_2 \rightarrow HBr + H$. This hydrogen atom then can react either with a bromine molecule: $H + Br_2 \rightarrow HBr + Br$; or with an HBr molecule, already formed: $H + HBr \rightarrow H_2 + Br$.

Note that HBr is formed and the chain is propagated by the two reactions $Br + H_2 \rightarrow HBr + H$ and $H + Br_2 \rightarrow HBr + Br$; each recurrence of this sequence converts one molecule of hydrogen and one of bromine to two molecules of hydrogen bromide and generates other atoms that reinitiate the sequence, thus continuing the chain. Finally, after most of the hydrogen and bromine molecules present have reacted, the chain-termination reaction, $Br + Br \rightarrow Br_2$, predominates, with the formation of trace amounts of bromine molecules. Here, the symbols used in the equations represent atoms and molecules rather than moles of atoms and molecules. Stoichiometric application is not as appropriate for these details as it is for the overall synthesis equation.

Substitution, elimination, addition. Substitution reactions are reactions in which a molecule is changed by the loss of one or more atoms and the gain of one or more other atoms that in a sense substitute for those that are lost. For example, chloroform, $CHCl_3$, reacts with antimony trifluoride, SbF_3, to form the useful compound monochlorodifluoromethane, $CHClF_2$. The incomplete, nonstoichiometric equation, emphasizing only the substitution, is $CHCl_3 + SbF_3 \rightarrow CHClF_2$. Two fluorine atoms are substituted for two chlorine atoms in this equation. The product substance undergoes a further reaction when heated strongly: $2CHClF_2 \rightarrow C_2F_4 + 2HCl$.

This reaction is an example of an elimination reaction; in this case, a hydrogen atom and a chlorine atom are eliminated as molecular hydrogen chloride, HCl. In the presence of hydrogen peroxide or other catalysts, this compound, tetrafluoroethylene, C_2F_4, polymerizes to form the well-known substance Teflon, $(CF_2)_n$, or $\ldots CF_2CF_2CF_2CF_2CF_2CF_2CF_2CF_2 \ldots$.

Hydrogenation reaction

Addition reactions, as the name implies, are reactions in which atoms are added to a molecule. If the added atoms are hydrogen atoms, the addition reaction is called a hydrogenation reaction. For example, many different vegetable oils can be hydrogenated. The product is a solid that can be used as shortening in the preparation of food. Oleic acid, $C_{18}H_{34}O_2$, serves as an example: $C_{18}H_{34}O_2 + H_2 \rightarrow C_{18}H_{36}O_2$. This reaction and the hydrogenation of other vegetable oils are usually carried out in the presence of a specific catalyst, finely divided porous nickel; for this process to be economically effective, the hydrogen must be under high pressure.

Oxidation–reduction reactions. Oxidation–reduction reactions form another class of important chemical reactions. For example, the generation of electric current by the use of a so-called dry cell or by a storage battery can occur only by means of an oxidation–reduction reaction that takes place within the battery. (The chemical reactions that take place inside any typical battery remain unknown in detail.) In the more recently developed mercury cell, constructed out of zinc (Zn) metal, usually in the form of a cuplike container, in which mercury monoxide, HgO, water, and other substances are enclosed, in the region near the electrode marked with a plus sign, a reduction reaction occurs (the electron itself is represented by the symbol e^-, which can be taken to indicate [stoichiometrically] a mole of electrons): $2e^- + HgO + H_2O \rightarrow Hg + 2OH^-$.

This equation is called a reduction half-equation because it symbolizes the gain of electrons by the reagents, mercury monoxide and water, to form the products, mercury metal, Hg, and hydroxide ion, OH^-. At the corresponding negative electrode, an oxidation reaction occurs, electrons being lost during the formation of product from the reagent. The reaction is symbolized in the oxidation half-equation $Zn + 2OH^- \rightarrow ZnO + H_2O + 2e^-$.

In addition to the disappearance of hydroxide (OH^-) ion, zinc is also consumed. Hence, the cell will certainly cease to function when the enclosing cup, made of zinc, has disappeared. Further, the reduction half-equation states that, as this cell is discharged, metallic mercury is produced within the cell.

Acid–base reaction. Although the classes of chemical reaction have by no means been exhausted, the final class to be mentioned here is acid–base reaction. One definition of an acid is that it is a substance that gives up a proton; *i.e.,* a hydrogen ion, H^+ (a hydrogen atom with its single electron removed). Vinegar is largely composed of water and acetic acid, CH_3COOH. The hydrogen atom on the right end, as depicted here, can be lost as an ion with the electron remaining behind on the acetate residue, now to be identified as an acetate ion $(CH_3COO)^-$. $CH_3COOH \rightarrow (CH_3COOO)^- + H^+$.

The ionization of acetic acid, however, will not occur unless a base is present. A base is defined as a substance that tends to take protons from acids. Water is a suitable base, taking the proton to become a hydronium ion, H_3O^+: $CH_3COOH + H_2O \rightarrow (CH_3COO)^- + H_3O)^+$.

In water solution, sodium bicarbonate, $NaHCO_3$ (common baking soda), forms dispersed sodium ions, Na^+, and bicarbonate ions $(HCO_3)^-$, for the most part. Bicarbonate ions are bases, and they tend to take hydrogen ions from hydronium ions, which act as an acid, to form carbonic acid, H_2CO_3: $(H_3O)^+ + (HCO_3)^- \rightarrow H_2O + H_2CO_3$. The carbonic acid is unstable at ordinary temperatures; most of it decomposes into water and gaseous carbon dioxide: $H_2CO_3 \rightarrow H_2O + CO_2$.

SEPARATION AND ANALYSIS

Broadly, the techniques used to investigate chemical reactions often require that the products of the reaction first be separated, because they are usually formed together within the same enclosure, such as a test tube. Separation techniques often involve the precipitation of one of the products; *i.e.,* by means of a second chemical reaction, a product of the first is altered by addition of or elimination of one or more atoms, so that this new substance can be separated, perhaps by the addition of a liquid in which only it is soluble or in other ways.

Identification of products

Following separation, the product is identified by its properties. For example, water can be identified by its freezing point or its refractive index (a measure of the ratio of the velocities of light when it passes through one medium compared to another) or its elemental composition of 8 to 1 oxygen to hydrogen mass ratio or by the energy of particular photons that it absorbs when exposed to photons of different energies.

An example of a simple case indicates in a general sense a procedure that might be followed. If the amount of acetic acid present in a particular brand of vinegar is to be measured, one first prepares a solution of a base of known concentration, such as baking soda. Then, a measured volume of vinegar is put into a suitable reaction vessel, and, drop by counted drop, the solution of base is added. The volume of base solution added, including the last drop that caused the evolution of gas (carbon dioxide), would serve as a measure (in an approximate way) of the amount of acetic acid in the measured volume of vinegar originally put into the reaction vessel. For more precise work, a solution of hydroxide ion base, made by dissolving sodium hydroxide, NaOH, in water at a known concentration, would be used. In this instance, a few drops of coloured indicator solution, such as the red-coloured liquid made by steeping red cabbage in hot water, would be added to the vinegar before adding the solution of base. The chemical reaction is complete when the indicator, which is dispersed throughout the liquid, changes colour upon the addition of one drop more of base solution. Of course, the procedures currently used to elucidate the details of chemical reactions are considerably more sophisticated, but the principle is the same.

(J.A.Y.)

Mechanisms and kinetics

The mechanisms of chemical reactions are the detailed processes by which chemical substances are transformed into other substances. The reactions themselves may involve the interactions of atoms, molecules, ions, electrons, and free radicals, and they may take place in gases, liquids, or solids—or at interfaces between any of these.

Reasons for study

The study of the detailed processes of reaction mechanisms is important for many reasons, including the help it gives in understanding and controlling chemical reactions. Many reactions of great commercial importance can proceed by more than one reaction path; knowledge of the reaction mechanisms involved may make it possible to choose reaction conditions favouring one path over another, thereby giving maximum amounts of desired products and minimum amounts of undesired products. Furthermore, on the basis of reaction mechanisms, it is sometimes possible to find correlations among systems not otherwise obviously related. The ability to draw such analogies frequently makes it possible to predict the course of untried reactions. Finally, detailed information about reaction mechanisms permits unification and understanding of large bodies of otherwise unrelated phenomena, a matter of great importance in the theory and practice of chemistry.

Generally, the chemical reactions whose mechanisms are of interest to chemists are those that occur in solution and involve the breaking and reforming of covalent bonds between atoms—covalent bonds being those in which electrons are shared between atoms. Interest in these reactions is especially great because they are the reactions by which such materials as plastics, dyes, synthetic fibres, and medicinal agents are prepared, and because most of the biochemical reactions of living systems are of this type. In addition, reactions of this kind generally occur in time scales convenient for study, neither too fast nor too slow, and under conditions that are easily manipulated for experimental purposes. Lastly, there are a number of techniques by which the mechanisms of such reactions can be investigated.

Chemical reactions involve changes in bonding patterns of molecules—that is, changes in the relative positions of atoms in and among molecules, as well as shifts in the electrons that hold the atoms together in chemical bonds. Reaction mechanisms, therefore, must include descriptions of these movements with regard to spatial change and also with regard to time. The overall route of change is called the course of the reaction, and the detailed process by which the change occurs is referred to as the reaction path or pathway.

Also important to the study of reaction mechanisms are the energy requirements of the reactions. Most reactions of mechanistic interest are activated processes—that is, processes that must have a supply of energy before they can occur. The energy is consumed in carrying the starting material of the reaction over an energy barrier. This process occurs when the starting material absorbs energy and is converted to an activated complex or transition state. The activated complex then proceeds to furnish the product of the reaction without further input of energy—often, in fact, with a release of energy. Such considerations are important to an understanding of reaction mechanisms because the actual course that any reaction follows is the one that requires the least energy of activation. This reaction course is not always the one that would seem simplest to the chemist without detailed study of the different possible mechanisms.

Difficulties in studying reactions

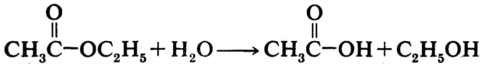
The study of reaction mechanisms is complicated by the reversibility of most reactions (the tendency of the reaction products to revert to the starting materials) and by the existence of competing reactions (reactions that convert the starting material to other than the desired products). Another complicating factor is the fact that many reactions occur in stages in which intermediate products (intermediates) are formed and then converted by further reactions to the final products. In examining chemical reactions it is useful to consider several general subjects: (1) factors that influence the course of chemical reactions, (2) energy changes involved in the course of a typical reaction, (3) factors that reveal the mechanism of a reaction, and (4) the classification of reaction mechanisms. With this information in mind it is then possible to look briefly at some of the more important classes of reaction mechanisms. (The sections *Acid–base reactions and equilibria, Oxidation–reduction reactions,* and *Electrochemical reactions* below deal with the mechanisms of reactions not described in this section.)

GENERAL CONSIDERATIONS

Identification of constituents

Determinants of the course of reaction. *The reactants.* In analyzing the mechanism of a reaction, account must be taken of all the factors that influence its course. After the bulk chemical constituents have been identified by ordinary methods of structure-determination and analysis, any prereaction changes involving the reactants, either individually or together, must be investigated. Thus, in the cleavage of the substance ethyl acetate by water (hydrolysis), the actual reagent that attacks the ethyl acetate molecule may be the water molecule itself, or it may be the hydroxide ion (OH^-) produced from it (see below).

The hydrolysis of ethyl acetate can be represented by the following equation:



$$CH_3\overset{\overset{\displaystyle O}{\|}}{C}{-}OC_2H_5 + H_2O \longrightarrow CH_3\overset{\overset{\displaystyle O}{\|}}{C}{-}OH + C_2H_5OH$$

ethyl acetate water acetic acid ethyl alcohol

in which the structures of the molecules are represented schematically by their structural formulas. An arrow is used to indicate the reaction, with the formulas for the starting materials on the left and those of the products on the right. In the structural formulas,the atoms of the elements are represented by their chemical symbols (C for carbon, H for hydrogen, O for oxygen), and the numbers of the atoms in particular groups are designated by numeral subscripts. The chemical bonds of greatest interest are represented by short lines between the symbols of the atoms connected by the bonds.

Important to this reaction is an equilibrium involving the cleavage of the water molecules into positively and negatively charged particles (ions), as follows:

$$2H_2O \rightleftharpoons H_3O^+ + OH^-$$

water hydronium ion hydroxide ion

In this equation, the numeral in front of the symbol for the water molecule indicates the number of molecules involved in the reaction. The composite arrow indicates that the reaction can proceed in either direction, starting material being converted to products and vice versa. In practice, both reactions occur together and a balance, or equilibrium, of starting materials and products is set up. The significance of this equilibrium for the hydrolysis of ethyl acetate is that any of the three entities (water molecules, hydronium, or hydroxide ions) may be involved in the reaction, and the mechanism is not known until it is established which of these is the actual participant. This often can be established if it is possible to determine the relative amounts of the three in the reaction medium and if it can be shown that the rate of the reaction depends upon the amount (or concentration) of one of them. Under certain conditions the hydrolysis of ethyl acetate is found to involve water molecules (as shown in the equation above); in other cases, hydroxide ion is involved.

The transition state. The transition state, or activated complex, is the fleeting molecular configuration that exists at the top of the energy barrier that the reactants must surmount to become the products. Strictly it is not a component of the reaction system and it cannot be examined directly in the way that an intermediate (however unstable) can because it lasts no longer than the duration of a molecular collision. The transition state may have properties of its own, not reflected in those of the starting materials or of the products and of the reaction, and so it is of vital importance in determining the course of reaction. Inference concerning the nature of the transition state is the essence of mechanistic study.

The solvent. The solvent, or medium in which the reaction occurs, may perform the mechanical, but often vital, role of allowing otherwise immiscible reactants to come together rapidly. Among the important groups of solvents, each with its own special type of behaviour, are hydroxylic solvents (those the molecules of which contain hydroxyl [-OH] groups, such as water and alcohols), dipolar aprotic solvents (those the molecules of which show a separation of electrical charge but do not easily give up a proton, or positive hydrogen ion; *e.g.,* acetone), and nonpolar solvents (those the molecules of which do not show charge separation; *e.g.,* hexane).

Solvent interactions

In dissolving the reactants, the solvent may interact with any or all of them, and it may be involved in the transition state for any reaction available for the system. If the solvent interacts more powerfully with the transition state than with the reactants, it facilitates the reaction. The solvent itself, of course, may be one of the reactants, and this circumstance introduces special problems because of the difficulty of distinguishing experimentally between its functions as a reagent and as an environment for the reaction.

Catalysts. Catalysts are substances that speed up a reaction by facilitating a particular mechanism—sometimes by influencing an existing prereaction and sometimes by making a new process energetically favourable. Their presence or absence frequently determines the course a reaction may take, simply because one of a number of competing reactions is, or is not, favoured. (Most catalysts are changed chemically while they speed up a reaction: sometimes—but not always—they are consumed; sometimes they are reformed and so appear to be unchanged in concentration during a reaction.)

The products. All reactions are reversible in principle, and the nature of the products of the reaction can affect the reaction course in a number of ways. When the position of equilibrium is unfavourable, for example, the accumulation of products may cause a reversal of the reaction. In such circumstances, the physical removal of the products (either through their volatility or insolubility, for example) facilitates the completion of the forward process. Sometimes, too, one of the products acts as a catalyst or as an inhibitor, behaviour that strongly influences the course of the reaction.

The reaction conditions. The conditions under which some reaction occurs, including such variables as the temperature and concentrations of reactants, also are important in determining the course of the reaction. For reactions that have a high energy barrier between reactants and products, the rate is highly responsive to change in temperature, and such reactions become more likely at increased temperatures, so that the minor products of a reaction often appear in larger proportion at higher temperature.

Similarly, the concentration of reagents can be important to the course of a reaction, especially if two mechanisms are available that involve different numbers of molecules in the transition states. Higher concentrations of a particular reagent favour those mechanisms in which greater numbers of molecules are involved in the transition state.

Importance of concentration of reagents

The pressure applied to the reacting system also may be significant; partly because it has an effect on concentration, and partly because mechanisms involving closely associated transition states become more favourable at high pressures. The latter relationship comes about because associated transition states are those in which several molecules or ions are brought close together (and therefore take up less space), a situation that is encouraged by increased pressures.

Energy changes involved in reactions. Collisions between molecules are rapid; therefore, reactions that occur spontaneously whenever the reagents collide are fast at ordinary concentrations. A reaction may be restricted in rate by its dependence on the occurrence of molecular collisions, however, because, for example, the reagents are present in such small amounts that reactions can only occur when they happen to encounter one another. Such a reaction is said to be diffusion-controlled because it is dependent on the process of diffusion to bring the molecules together. In such cases, the viscosity of the medium is relevant; the more viscous, or "thick," the medium, the more difficult the diffusion and the slower the reaction.

As noted above, however, most reactions involve a rate-limiting energy barrier, and it is the nature of this barrier and of the molecular configuration at its top that determines the mechanism. Diagrams of energy changes during the course of reaction often are used to illustrate the energetic aspects of the reaction. An example of a possible energy diagram for a hypothetical one-stage process, the dissociation in a solution of a covalent molecule designated E–N, into its ions, E^+ and N^-, is shown in Figure 1.

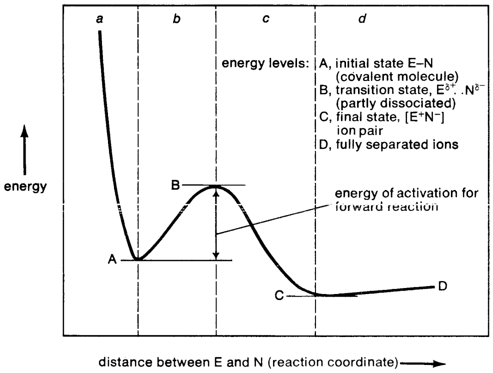


Figure 1: Possible energy diagram for the dissociation of a covalent molecule, E–N, into its ions E^+ and N^- (see text).

In this diagram the energy is plotted against a reaction coordinate—a spatial relationship that varies smoothly during the course of the reaction—in this case the distance between the portions of the molecule designated E and N. At the left-hand energy minimum of Figure 1, the bond between E and N is fully formed; if energy is applied to excite the system in such a way that E and N are brought even closer together (region *a* of Figure 1), the atomic nuclei repel, with the result that energy rises steeply. Alternatively, excitation energy, such as thermal energy from collisions with other molecules, may stretch the bond, and the energy curve then moves into region *b*. The E–N bond is thereby weakened steadily until the transition state is reached. This point, as can be seen in Figure 1, has the maximum energy on the reaction coordinate. At the same time, this point represents the minimum energy required to convert the reactants into the products. The curve shown should be considered as only a planar section of a three-dimensional energy surface relating to the various possible spatial relations among the components of the reaction. The passage of the reactants from the initial state to the products then can be thought of as analogous to the climb from a valley (the initial state) through the lowest mountain pass leading to a second valley (the products). Thus, although the transition state represents a peak on the single curve depicted, it really represents a secondary minimum (or pass) in the energy surface. From the top of this pass (the transition state), the molecule can only descend, losing energy by collision. In doing so it may revert to the starting materials, or it may dissociate to give the products (region *c* in Figure 1). The products in the case of the reaction chosen are the ions E^+ and N^-, which are held together by electrostatic attraction as an ion pair at the right-hand minimum of the graph; beyond this point, further separation of the ions involves the consumption of energy (region *d*). In principle, the products may lie at higher or (as shown) at lower energy than that of the initial state. The mechanism of a reaction such as this may be considered to be completely defined when the structures and energy properties of the starting materials, the products, and the transition state are known.

Computation from atomic structures

Ultimately, it should become possible to compute the properties of the molecules solely from the properties of their constituent atoms and also to deduce the transition states for any of the reactions these molecules may undergo. For a few simple situations, approaches already have been made to definitions of mechanism in this degree of detail. Systems involving several atoms, however, require a many-dimensional representation of the reaction course instead of the two-dimensional description shown in Figure 1. The problems of computation, and of testing theory against experiment, then become enormous. Nonetheless, attempts have been made to deal with some simple reactions of systems involving up to about five atoms.

Multistage. For a reaction involving several distinct stages, a more complicated description of the reaction course also is necessary. Figure 2 gives an example of such a situation. This hypothetical reaction is reversible, with three successive intermediate complexes formed between the reactants. Unlike the case of the simpler situation above, the physical process that best approximates what is happening along the reaction coordinate changes from stage to stage across the diagram.

Principle of microscopic reversibility

The highest point on the energy diagram (Figure 2) corresponds to the energy of the transition state of the rate-limiting step in the reaction—that is, the slowest step in the reaction, the one that governs, or limits, the overall rate. The rate of reaction is independent of the nature and number of the intermediates that lie before this transition state on the reaction coordinate. Progress along the reaction coordinate cannot be identified with the time course of the reaction because any individual pair of reactants may reside for an appreciable time in the partially activated state represented by one of the intermediate complexes before final reaction is achieved. Once the reactants have passed the rate-limiting transition state they must lose energy (usually by collision with other molecules) to reach the final state. In a sense the reaction coordinate in such a reaction may be thought of as representing the chemical course of the reaction (rather than a spatial or a time course). If the diagram represents the only reaction of the system, then it is possible to apply the principle of microscopic reversibility, which states that the course taken by reverse action will be statistically identical with that taken by the forward reaction. This principle is not, however, applicable to a reaction giving several different products not at equilibrium with one another.

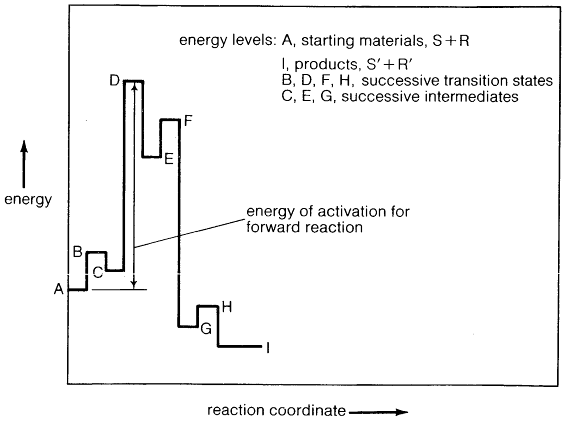


Figure 2: Energy levels in a hypothetical multistage reaction (see text).

Identification of a reaction mechanism. *Chemical and stereochemical nature of reactants, intermediates, and products.* The stoichiometry of a reaction consists of the chemical formulas and relative molecular proportions of starting materials and products. Obviously these have a bearing on the mechanism of the reaction, for the overall reaction course must proceed from starting materials to the products. The stoichiometry of the reaction may be misleading, however, because the participants in the overall reaction may not be involved directly in the rate-limiting step.

The discovery of intermediates in the course of a reaction is important because these point to the existence of distinct stages, the mechanism of each of which must then be determined. The identification of intermediates that persist only briefly or that are present in only small amounts depends on the availability of powerful, sensitive, and rapid experimental techniques. For this purpose, a number of specialized instrumental procedures (including ultraviolet, infrared, magnetic resonance, and mass spectrometry) are widely used to supplement the more usual chemical and physical methods.

The identification of a new chemical substance formed transiently in a reaction mixture, however, does not unambiguously imply that that substance is an intermediate in the reaction. In many cases a newly found material is only a temporary repository of the proportion of the reactants and ultimately produces the products by first reverting to the starting material.

Identification of products

The identification of the products of a reaction also helps to define the reaction course, because the mechanism in question clearly must account for their formation. Mechanistic theory has been greatly facilitated by the development of powerful methods of separation and purification based on chromatography (separation of compounds on the basis of their relative degrees of adsorption to certain solid substances, such as starch or silica) and also by modern methods available for the analysis of small quantities of materials. These spectroscopic procedures are often used, as is another instrumental method, called polarimetry.

An important consideration with regard to the products of the reaction is whether the reaction is under kinetic or thermodynamic control. A reaction is said to be kinetically, rather than thermodynamically, controlled, when the products are formed in proportions different from those that would prevail at equilibrium between the same products under the same conditions. Thermodynamic control leads to the equilibrium ratio of the products. Often, though not invariably, reactions under kinetic control give a greater amount of the thermodynamically less stable of two possible products; if thermodynamic control is then established, the products shift to their equilibrium proportions, which might give a misleading picture of the reaction course. Hence inferences concerning the nature of the transition state can be drawn from the nature of the products only with a good deal of circumspection.

In determining the mechanism of a reaction, one of the major problems is to deduce the spatial or three-dimensional changes that occur to the molecules involved as they proceed from their initial state through the intermediate stages and transition states to the final products. Knowledge about such changes generally can be deduced from knowledge of the stereochemistry (three-dimensional structures) of the starting materials, intermediates, and final products (provided these are obtained under kinetic control). Information of this kind is obtained by determinations of optical activity and analysis of the structures of the compounds by standard means.

In certain instances information about the movement of atoms between molecules during the course of a reaction can be gained by using compounds containing isotopes of certain of the atoms. These isotopes behave much like the ordinary atoms they replace, but they can be identified by their behaviour. For example, in the hydrolysis of ethyl acetate (see above *The reactants*), it is crucial to a determination of the mechanism to be able to establish which of the two reactants (ethyl acetate or water) provides the oxygen atom that ends up in the product ethyl alcohol. In this case, the use of water labelled with oxygen-18 reveals that the oxygen atom in the alcohol comes from the ethyl acetate molecule.

Kinetic order. The possibilities that need to be considered for the transition state having been limited by determination of the chemical structures of the participants, the most powerful method of obtaining further information is the use of the kinetic method; *i.e.,* the study of the effect of reaction conditions on the rate of reaction. Experimental methods that have been used in kinetic studies include most of the known methods of chemical separation and analysis. Techniques that involve removing samples from the reaction mixture at intervals or stopping the reaction and analyzing for starting material or product are common for reactions with half-lives down to about a minute. For faster reactions, methods involving rapid scanning and automatic recording of some characteristic property of the reacting mixture, such as absorption of light at a particular wavelength, recently have become important. Other procedures for following exceptionally fast reactions include the controlled supply of a reagent within extremely small concentration limits, sometimes by electrolytic procedures (the use of an electric current to produce precise amounts of the substance), and sometimes by carrying out the reaction under conditions in which separate flowing streams of the reactants come together to insure rapid mixing. So-called relaxation procedures, in which a system in equilibrium is very rapidly perturbed and its rate of relaxation to the original or to some new equilibrium state is observed, also have been applied to the study of reactions of extremely small half-lives.

Effects on reaction rates

The mechanistic information to be obtained from the observed kinetic behaviour of a system derives from the fact that, for an activated process, the transition state can be considered to be in thermodynamic equilibrium with the starting materials except with respect to its motion along the reaction coordinate. It follows that the rate of reaction is approximately proportional to the product of the concentrations of those substances that comprise the transition state. If the concentrations of all but one reactant are held constant while the concentrations of that reactant are changed, then the variation in rate with the concentration changes will establish how many molecules of that particular reactant are involved in the transition state. This figure is called order of reaction with respect to the reactant in question. A full description of the composition of the transition state then requires identification of the order of reaction with respect to each of the reactants.

Although straightforward in principle, the application of this method involves some difficulties. Sometimes (for example, with ionic solutes in solvents of low polarity), the concentrations of the reagents are not truly representative of their influence on the reaction rate. The kinetic order then does not properly represent the composition of the transition state. The power of the kinetic method is often greatly increased when the rate of reaction can be followed by more than one method. In such instances, it frequently has been found that unexpected differences reveal the intervention of previously unsuspected intermediates.

Environmental effects. Changes in the environment (such as the composition of the solvent) frequently influence the course of the reaction by affecting the relative stabilization of the initial state and the transition state. Large changes in the polar character (charge distribution) of the solvent, for example, may have an effect on the course of the reaction if there is a substantial change in polarity between the reactants and the transition state.

Correlation of rate with structural change

Structure–reactivity relations. An observed correlation of changes in a reaction rate with systematic changes in the structure of one of the reactants often reveals the movements of electrons between atoms as the reactants shift toward the transition state. Systematic changes in structure usually are brought about by selecting a particular molecular system and varying a portion of it (such as, for example, the substituents on a benzene ring). The effects of each variant on the rates of several different reactions are determined experimentally, and the results plotted on graphs, the values resulting from a particular molecular variation in one reaction being measured along one axis and those of the same variation in the other reaction along the other axis. A straight-line relationship indicates that the molecular changes are affecting the rates of the two reactions in related ways. The slope of the line gives a comparison of the relative response of the two systems to the given change in structure; and the sign of the slope tells whether a particular structural change favours both reactions, or favours one while disfavouring the other. The observed effects generally can be correlated with the electronic nature of the molecular variants introduced. For example, if a substituent in the molecule tends to donate electrons toward the reactive centre in the molecule, and this change favours the reaction, it can be concluded that an electron-rich centre is involved in the transition state.

Electronic effects of the above kinds can be complicated by spatial, or steric factors. (A reaction is said to be sterically hindered when the transition state is more congested than the initial state and to be sterically accelerated when the reverse is true.) When suitable allowance is made for the above electronic influences, structural changes can be used to help define the detailed geometry of the transition state. Thus if large or bulky substituents have an inhibiting effect on the course of the reaction, it can be concluded that the transition state differs from the starting material in such a way that the effect of the bulky group is accentuated; in the alternative situation, in which a bulky substituent accelerates the reaction, it may be concluded that the formation of the transition state relieves crowding found in the starting material. Although absolute calculations of reactivity—that is, calculations based on molecular structure alone—have made little progress in the case of polyatomic systems, significant calculations of reactivity differences have been made in favourable instances in which steric and electronic effects can be disentangled.

Carbon–hydrogen and carbon–deuterium bonds

Kinetic isotope effects. Isotopes are atoms that have the same atomic number (and, hence, generally the same chemistry) but different mass. The difference in mass becomes chemically important in certain instances. For example, when a carbon–hydrogen bond is replaced by a carbon–deuterium bond (deuterium being an isotope of hydrogen with about twice the mass), the vibrational frequencies of that bond are changed. The vibrational stretching frequency, for example, of a bond between two atoms gives an approximate measure of the bonding forces holding those two atoms together, the effective masses of the two atoms being allowed for. If the character of the carbon–hydrogen bond is altered between the normal state and the transition state, the change from hydrogen to deuterium may have an effect on the relative stabilities of the normal and the transition states, and also, therefore, an effect on the rate of reaction. Such effects, called kinetic isotope effects, operate when only one bond is concerned in such a way that bonds involving the heavier isotope are broken with more difficulty than those involving the lighter isotope. Isotope effects are large only for the isotopes of hydrogen; but, with heavier elements, even small differences can give important information about the mechanism, provided that sufficiently precise methods are available for their measurement.

Classification of reaction mechanisms. There is no one generally agreed and completely satisfactory method of classifying mechanisms; individual authors have often adopted their own nomenclature and symbolism. There are, however, a number of useful classification principles that should be noted.

Homolysis and heterolysis. When a covalent bond (a nonionic chemical bond formed by shared electrons) is made up of two electrons, each of which is supplied by a different atom, the process is called colligation; the reverse process, in which the electrons of a covalent bond are split between two atoms, is known as homolysis. These reactions are shown schematically by the equation:

$$A\cdot + B\cdot \underset{\text{homolysis}}{\overset{\text{colligation}}{\rightleftarrows}} A:B$$

in which A and B represent the separate atoms (or groups), the single dots represent electrons, and the double dots represent the electron pair that makes up the bond. The products of a homolysis reaction are called free radicals, and all such processes are said to have homolytic or free-radical mechanisms.

Covalent bond breaking

If, on the other hand, a covalent bond is formed by a pair of electrons, both of which come from one of the two reagents, the process may be described as coordination, and its reverse is heterolysis. Coordination and heterolysis are shown schematically by the equation:

$$N\!: + E \underset{\text{heterolysis}}{\overset{\text{coordination}}{\rightleftarrows}} N:E$$

in which the dots indicate the electron pair and the letters N and E represent the atoms (or groups) that, respectively, donate and accept the electrons (see below for special significance of the letters N and E). Reactions of this kind are said to have heterolytic mechanisms.

Nucleophilicity and electrophilicity. In a heterolytic reaction, the unit that carries the electron pair (designated N: above) is nucleophilic; *i.e.,* it seeks an atomic nucleus to combine with. Conversely, the other unit in the reaction (designated E) is electrophilic: it seeks to combine with a pair of electrons. An electrophilic reaction mechanism is one that involves an electrophilic reagent attacking a nucleophilic substrate. Because every such reaction involves both an electrophilic substance and a nucleophilic substance, it must be agreed arbitrarily which unit is the reagent and which the substrate. Often this agreement is made on the basis of molecular size, the larger sized material being classed as the substrate.

In certain reactions, in which movements of electrons are concerted and cyclic, it is not possible to identify any one reagent, or even any one particular atom or set of atoms in the molecule, as electrophilic or nucleophilic. Such processes are classified as electrocyclic.

Molecularity. The mechanism of an individual stage of a reaction can be described as unimolecular, bimolecular, and so on, according to the number of molecules necessarily concerned in covalency change in the transition state. As an extension of this classification, the number of molecules involved in the rate-limiting step of a several-stage reaction also can be used for classification of the overall reaction. Ambiguities and blurred distinctions arise when there are strong interactions between the solvent and the initial or transition state, or when two stages of a reaction are so near in rate that they become jointly rate-determining. Nevertheless, classifications based on molecularity are widely used.

Intermolecularity and intramolecularity. The distinction between intermolecular and intramolecular processes is often useful. In intermolecular reactions, covalency changes take place in two separate molecules; in intramolecular reactions, two or more reaction sites within the same molecule are involved.

Nature of catalysis. Classification also can be made on the basis of the mode of catalytic action. In ester hydrolysis, as the hydrolysis of ethyl acetate (above), for example, distinction can be made between mechanisms in which catalysis is brought about by protons (hydrogen ions), or by acids in general; by hydroxide ion, or by bases in general; or finally by enzymes.

Kinetic order. The kinetic order of a reaction is best considered as an experimental quantity related to (but not identical with) the number of molecules of any reactant involved in the transition state. It may for various reasons reveal only a part of what is happening in the rate-limiting transition state, one reason being that the concentrations of the components of the transition state may not all change with progress of the reaction. Establishing that a reaction is of the first order kinetically (that is, behaves as though only one molecule of the reactant is involved in the transition state) with respect to one of the components, for example, seldom reveals whether or not one or more molecules of solvent also is involved in the transition state. This is so because the solvent is present in such large amounts that its concentration does not change effectively with the course of the reaction. For this reason, the order with respect to an individual component may sometimes be useful for classification, but for the overall reaction molecularity is the more fundamental quantity.

Importance of molecularity

Time sequence of events. The time sequence of events in a chemical reaction also provides a means of classification. In some mechanisms, the bond-making and bond-breaking processes occur together and are said to be concerted; in others, the individual stages are discrete, with recognizable intermediates occurring between them, and it may be necessary to specify not only that the mechanism is stepwise but also the order in which the steps occur.

COMPARISON OF SELECTED REACTION MECHANISMS

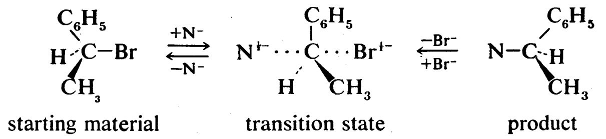
For the following incomplete and abbreviated survey of reaction mechanisms, several mechanisms important in the development of mechanistic study have been chosen.

Nucleophilic substitutions at saturated carbon centres. The term substitution refers in general to the replacement of any group in a molecule by any other group. Saturated carbon centres are carbon atoms at which no multiple bonds occur, and nucleophilic substitutions—those brought about by nucleus-seeking reagents—can occur at such carbon atoms by either of two main mechanisms, bimolecular and unimolecular.

Bimolecular. In bimolecular nucleophilic substitution reactions in which the substrate is attacked at a saturated carbon atom, the starting material has a tetrahedral structure and the transition state has a trigonal bipyramidal structure (both of which are shown below). Each individual act of substitution produces a product of inverted (*i.e.,* mirror-image) stereochemical configuration.

Inversion of structure

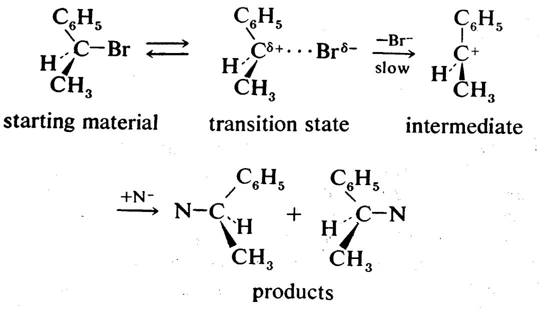
A typical bimolecular substitution reaction is shown by the equation:



in which the chemical symbols represent atoms of the elements as above (with Br the symbol for an atom of bromine and N the symbol for any nucleophilic agent). This equation differs from the earlier ones in that a three-dimensional representation of the structures is intended. The three-dimensional effect is achieved by considering that the bonds represented by ordinary solid lines lie in the plane of the paper, bonds represented by dashed lines project to the rear, and bonds represented by dark triangles project to the front. A further unique feature of this equation is that the representation of the transition state shows half bonds (bonds in the process of being formed or broken), which are indicated by dotted lines. In addition, in the transition state, half negative charges are indicated by the symbols "½–." The mechanism of this reaction is characterized by entry of the nucleophilic reagent from one side of the substrate molecule and departure of the bromide ion from the other side. The resulting change in configuration of the substrate has been likened to the turning inside out of an umbrella, with the transition state representing that precise moment when the ribs are essentially vertical in the course of their passage from one side of the structure to the other. The reaction is synchronized, or synchronous, in that entry of the nucleophile and departure of the leaving group occur simultaneously. It is bimolecular in that one molecule each of substrate and nucleophile are involved in the transition state; and it is stereospecific in that the stereochemical outcome of the reaction is invariably the same.

This bimolecular mechanism occurs with a wide range of structures. It often can be characterized by second-order kinetics, *i.e.,* by reaction rates that are dependent on the concentrations of both the substrate and the nucleophilic reagent. The transition state is highly congested, so that effects of steric hindrance are large. Otherwise, however, structural changes produce a variable response because of the conflicting electronic requirements of the bond-forming and bond-breaking processes. Bimolecular nucleophilic substitutions with rearrangement of the bonding skeleton also are known.

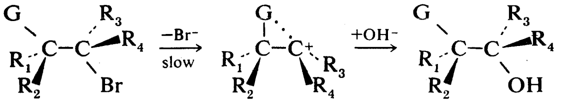
Unimolecular. Unimolecular nucleophilic substitution reactions proceed by a two-stage mechanism, in which heterolysis precedes reaction with the nucleophile. The following equation is a typical example:



in which the symbols are the same as in earlier equations, with the addition of delta plus (δ+) and delta minus (δ–), which indicate partial positive and negative charges respectively. The significant consideration in this reaction mechanism is the initial separation of the bromide ion (by way of a transition state showing partial separation of the ion) to give a free positively charged organic ion (carbonium ion). This step is the rate-determining step of the reaction and, because it involves only a molecule of the substrate, the reaction is unimolecular. The second stage of the reaction is the interaction of the intermediate carbonium ion with the nucleophile to give the products of the reaction.

The unimolecular reaction is characterized experimentally by first-order kinetics; *i.e.,* by a rate that depends only on concentration of the substrate (and not the nucleophile); by the absence of effects of steric hindrance; by powerful facilitation of the reaction by the presence of electron-releasing groups attached to the reaction centre; and by variable, and often diagnostic, stereochemistry. Inversion of stereochemical configuration (change from one configuration to the mirror-image configuration) is frequently encountered, accompanied by racemization (production of both mirror images). The extent of racemization depends upon the life of the intermediate carbonium ion, with longer lived ions leading to more extensive racemization (due to the fact that the symmetrical ion is exposed to attack from either side).

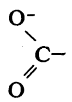
In an important group of structures, a group not formally involved in the overall reaction interacts with a carbonium ion centre to form an intermediate, which then reacts with the nucleophile to give a product of the same stereochemical configuration as the starting material. This behaviour can be represented by the equation:



G=participating or neighbouring group

R_1, R_2, R_3, and R_4=substituents

In the first demonstrations of this behaviour, the participating group (G) was a carboxylate anion group, which can be represented in chemical symbols as:



Neighbouring-group participation

Subsequent investigations have revealed numerous examples involving other substituents, and the phenomenon is now commonly described as neighbouring-group participation.

A frequent consequence of reaction through intermediates having carbonium ionic character is that some of the products have rearranged skeletal structures. In this equation, the symbol Cl represents a chlorine atom.

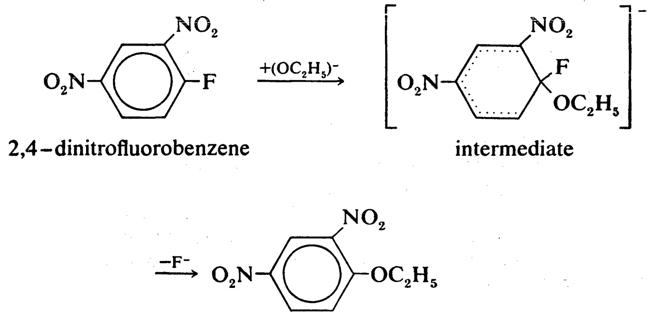
$$\underset{\text{1-chloro-2,2-dimethylpropane}}{(CH_3)_3C{-}CH_2Cl} \xrightarrow{-Cl^-} \underset{\text{intermediate carbonium ion}}{[(CH_3)_3C{-}CH_2^+]}$$

$$\longrightarrow \underset{\text{rearranged carbonium ion}}{[(CH_3)_2C^+{-}CH_2CH_3]} \xrightarrow{+OH^-} \underset{\textit{tert}\text{-amyl alcohol}}{(CH_3)_2C(OH){-}CH_2CH_3}$$

The fundamental difference between the transition states in the bimolecular and unimolecular mechanisms is the degree of covalent bonding between the nucleophile and the substrate in the transition state. In the unimolecular mechanism such bonding is negligible; in the bimolecular case, it has essentially reached the half-bond status. In borderline situations, the matter is difficult to resolve, a number of intermediate cases being known, and there has been much controversy as to the validity of the distinction between the bimolecular and the unimolecular mechanisms. Experimentally, however, clear examples of each class have been established.

Resistance to nucleophilic reagents

Nucleophilic substitution at unsaturated carbon centres. Unsaturated carbon centres—including those involving ordinary carbon–carbon double bonds and those involving the extended cyclic systems of alternate single and double bonds known as aromatic rings—are not easily attacked by nucleophilic reagents unless they have been denuded of electrons by electron-attracting substituents. A two-stage process that includes addition of the nucleophile followed by expulsion of a negatively charged (anionic) group is the course normally taken for substitutions at aromatic centres. The presence of the aromatic ring enforces the geometry of the product, and the reaction is favoured by electron-withdrawing groups, such as the nitro ($-NO_2$) group, which help to accommodate the negative charge on the intermediate. An example of this type of reaction is the displacement of fluoride ion from 2,4-dinitrofluorobenzene by nucleophiles such as ethoxide ion.



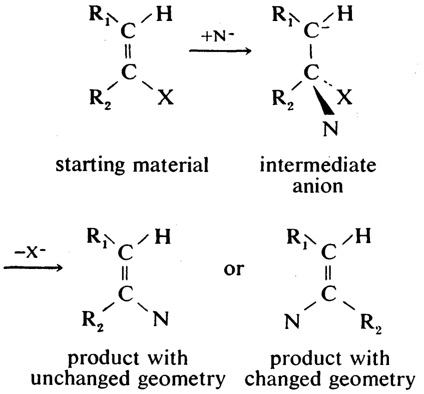
2,4-dinitrofluorobenzene

intermediate

2,4-dinitrophenyl ethyl ether

In this equation fluorine atoms are indicated by the chemical symbol F; nitro groups (consisting of one nitrogen and two oxygen atoms) are indicated by the symbols $-NO_2$; normal benzene rings (of six carbon atoms, each of which carries a single hydrogen atom) are indicated by regular hexagons with circles in them; and benzene rings containing disrupted electronic structures are indicated by hexagons with partial dotted circles.

Substitution reactions at ordinary double bonds (olefinic bonds) also take place by a two-stage process. When the two stages in the reaction occur synchronously or in very quick succession, the product has the same geometrical relationship that existed in the starting material. If, however, the anionic intermediate has sufficient lifetime, rotation about the new carbon–carbon single bond can precede loss of the negatively charged group, resulting in production of two products of differing molecular geometry—that is, products in which the substituents are differently situated with respect to the double bond.



starting material

intermediate anion

product with unchanged geometry

product with changed geometry

X = a halogen atom: fluorine, chlorine, bromine, or iodine

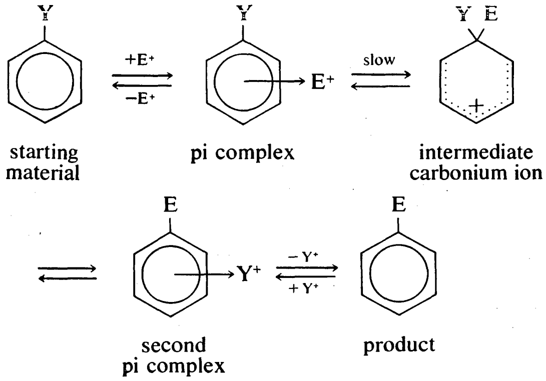
If the intermediate anion takes up a hydrogen ion (proton) and then loses hydrogen and halogen simultaneously (concerted elimination), the reaction then is said to be following an addition–elimination sequence. Examples of such reactions are known, particularly in situations in which the double bond includes an atom other than carbon. In aromatic systems, the reverse situation, in which elimination occurs, followed by addition, also is found. Finally, unimolecular mechanisms of substitution also are known to take place at particularly activated unsaturated centres. For example:

$$Ar-C(Br)=CH_2 \xrightarrow{-Br^-} Ar-C^+=CH_2 \xrightarrow{+OEt^-} Ar-C(OEt)=CH_2$$

in which the symbol Ar represents a benzene ring or other aromatic system.

In aromatic systems

Electrophilic substitution at unsaturated carbon centres. Because of its wide applicability, particularly to aromatic systems, electrophilic substitution is an important reaction. Reaction by any one of several mechanisms is possible. One of the more common is shown here; reactions in this category consist of replacement of a group designated Y (often a hydrogen atom) in an aromatic molecule by an electrophilic agent designated E. Both substituents can be any one of various groups (*e.g.*, hydrogen atoms, or nitro, bromo, or *tert*-alkyl groups).



starting material

pi complex

intermediate carbonium ion

second pi complex

product

Here, Y represents a substituent on the ring; the arrow from the ring centre indicates coordination.

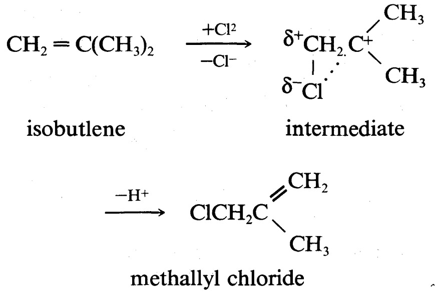
As shown, the reaction begins with formation of a pi complex, in which the electrons associated with the aromatic ring, or other unsaturated centre (pi electrons), coordinate weakly with the electrophile. This complex forms rapidly in an equilibrium preceding the rate-determining step, which itself leads to a carbonium ion intermediate and then by way of a second pi complex to the product. Examples are known in which the removal of the proton from the carbonium ion intermediate (to form the second pi complex) becomes rate-determining.

Reactivity by this mechanism is dominated by the electrophilic character of the reagent (E); it also, however, responds powerfully to changes in structure of the organic substrate. As would be expected, substituents that release electrons toward the reaction site facilitate the reaction, and those that withdraw electrons retard reaction. These effects are very specific with regard to the position at which the modifying group is introduced.

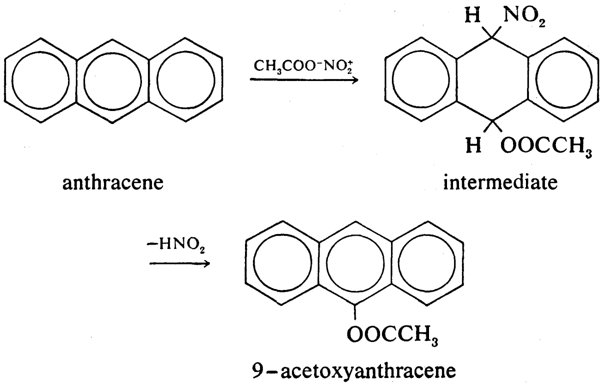
Steric (spatial) effects generally are smaller than electronic effects in determining the characteristics of reaction by this mechanism; but they are not negligible. Direct steric hindrance and steric acceleration both have been found with suitably placed large substituents and reagents, and indirect effects arising because one group interferes with the orienting power of another also are known.

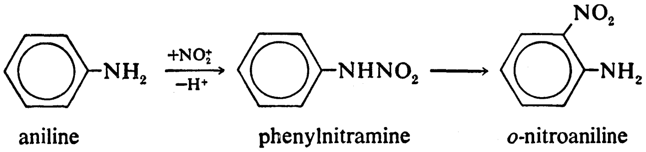
Substitutions with rearrangement

Substitution with accompanying rearrangement of the double-bond system is another established reaction path. An example is shown below in which the positions of chlorine attachment and proton loss were established by isotopic labelling.



Addition–elimination and indirect substitution reactions also can occur and are responsible for a number of unusual products formed in aromatic substitution reactions. Examples of these reaction sequences are shown below:





Addition reactions. Reactions in which a multiple bond between two atoms becomes partly or fully saturated by covalent attachments at both centres are called addition reactions. Many mechanisms are known for such reactions; most of them are variants of four basic mechanisms, which differ chiefly in the sequence of events that occur.

With initial electrophilic attack. Addition reactions beginning with electrophilic attack include many additions to olefins (compounds with double bonds), some additions to acetylenes (compounds with triple bonds), and some additions to compounds with other multiple bonds. There is a close relationship between this mode of addition and the electrophilic substitutions discussed in the preceding section, as shown by this general representation of the reaction:

$$\overrightarrow{CH_3}CH\overset{\frown}{=}CH_2 + H^+ \longrightarrow CH_3\overset{+}{C}HCH_3 + Cl^-$$

propylene proton intermediate carbonium ion chloride ion

(not $CH_3CH_2\overset{+}{C}H_2$)

$\longrightarrow CH_3CHClCH_3$;

2-chloropropane

(not $CH_3CH_2CH_2Cl$

1-chloropropane)

in which the arrows on the olefin structure indicate the flow of electrons toward the terminal carbon, which attracts the electrophilic proton because it becomes an electron-rich centre. Electrophiles, which can be effective either as positive ions (E^+) or in combination with a nucleophile (E–N), include protons (H^+), carbonium ions (R_3C^+), positively charged halogen ions (Cl^+, Br^+, I^+), nitronium ions (NO_2^+), nitrosonium ions (NO^+), and many others. In general, any nucleophile can complete the reaction. When the first stage of the reaction (addition of the electrophile) is rate-determining, the rate responds powerfully to electron-release to the reaction centre, and this factor determines selectively the orientation of initial attack with respect to the double bond. Thus propylene reacts with hydrogen chloride many times faster than ethylene does, and the product is exclusively 2-chloropropane, rather than 1-chloropropane, because the concentration of electrons on the terminal carbon determines that the electrophilic proton finds it easier to attack that carbon rather than the central carbon atom.

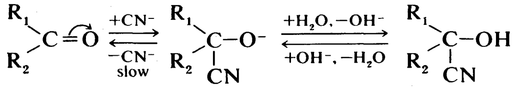
$$CH_2 = CH_2 + H^+ \longrightarrow CH_3\overset{+}{C}H_2 + Cl^- \longrightarrow CH_3CH_2Cl$$

ethylene proton intermediate carbonium ion chloride ion ethyl chloride

Addition by this mechanism can be accompanied by substitution and by rearrangement as alternative reactions of the carbonium ionic intermediate. Characteristically, the ratios of product are kinetically controlled (see above). Reactions by this mechanism can be complicated by the intervention of intermediates that are more complicated structurally. Neighbouring-group interaction can modify the structure of the intermediate toward a bridged structure and thus determine the stereochemistry of addition.

Although it is common to find that the first stage of this sequence is rate-determining, in some cases the rate-limiting transition state lies later along the reaction path. It also is possible for the two stages to be concerted, with the electrophilic and nucleophilic fragments (E and N) of the reagent E–N acting either as still covalently bound or as separate kinetic entities (E^+ and N^-). Especially in acid-catalyzed additions to carbon–oxygen and carbon–nitrogen double bonds, the first stage of the reaction can become rapidly reversible, and the mechanistic characteristics of the reaction are then appropriately modified.

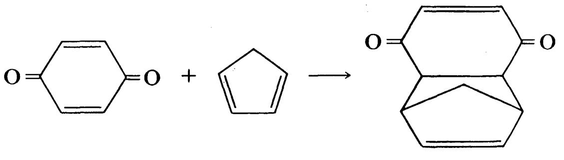
With initial nucleophilic attack. The reverse mode of addition, in which a nucleophile initiates attack on the multiply bonded carbon atom, is less easily realized in simple systems; but it does occur with acetylenes, and it also is the basis of reactions that occur when the centre of attack is denuded of electrons. For example, the formation of substances called cyanohydrins from carbonyl compounds (materials with carbon–oxygen double bonds) occurs as follows:



carbonyl compound intermediate cyanohydrin

in which the curved arrow indicates the movement of electrons in the carbonyl group. Initial attack on carbon by the nucleophilic cyanide ion in this case is facilitated by the electron withdrawal by the oxygen atom (shown by the curved arrow in the formula). Such electron withdrawal also can be transmitted along a series of alternate double and single bonds (a conjugated system), with resultant addition to the ends of the system.

Electrocyclic. In a third class of additions, both portions of the attacking reagent combine simultaneously with the substrate. Reactions of this kind sometimes retain predominantly electrophilic or predominantly nucleophilic character, as can be shown by structural and environmental effects. In a number of important cases, however, quite different behaviour is observed. For example, the addition of cyclopentadiene to 1,4-benzoquinone follows second-order kinetics and proceeds at nearly the same rate in the gas phase and in solvents of widely differing polar character.



1,4-benzoquinone cyclopentadiene adduct

In this equation the polygons represent rings of carbon atoms (one at each corner) with double bonds between certain atoms as shown. There must, therefore, be little development of charge in the transition state, and the formation of the two new single bonds and the accompanying electronic movements must be well synchronized. A large number of such reactions are known; they are characterized by a remarkable stereospecificity (stereochemical specificity), controlled in part by steric effects and in part by the stereo-electronic characteristics of the combining double-bond systems.

Synchronized reactions

Homolytic. Additions by free-radical mechanisms also are well known. They replace the concomitant polar additions most easily when homolytic (decomposition of a compound into two neutral atoms or radicals) fission of the reagent can be readily catalyzed and when the radicals produced as intermediates sustain chain processes. Addition of hydrogen bromide to olefins falls into this class. Equations (1)–(4) describe the main part of the sequence; reactions (2) and (3) are repeated many times before reaction (4) or some other reaction intervenes to break the chain. As a result, one act of initiation results in many molecules of product.

(1) $HBr \longrightarrow H\cdot + Br\cdot$ (chain-initiating step)

(2) $Br\cdot + CH_3CH = CH_2$
$\longrightarrow CH_3\dot{C}HCH_2Br$ (chain-propagating step)

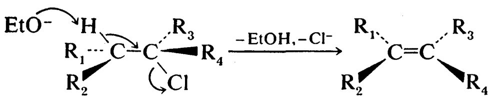
(3) $CH_3\dot{C}HCH_2Br + HBr$
$\longrightarrow CH_3CH_2CH_2Br + Br\cdot$ (chain-propagating step)

(4) $CH_3\dot{C}HCH_2Br + H\cdot$
$\longrightarrow CH_3CH_2CH_2Br$ (chain-terminating step)

The reaction can give an orientation of substituents opposite to that found in electrophilic addition [which in the above example would produce $CH_3CH(Br)CH_3$]; and in suitable cases it can be just as stereospecific.

Elimination reactions. Elimination reactions can be treated formally as the reverse of additions. The simplest examples of this class of reactions arethe olefin-forming 1,2-eliminations—that is, elimination of substituents from adjacent carbon atoms—but eliminations to give other types of double bonds are equally well known. Again, 1,3-eliminations—eliminations of substituents from carbon atoms separated by a third carbon—give compounds with three-membered rings of carbon atoms (cyclopropanes). Furthermore, the so-called conjugate eliminations occur when one or more double bonds are inserted between carbon atoms bearing the substituents that are eliminated; the result of such eliminations is a system of alternating double and single bonds (a conjugated system). Finally, there also are fragmentation reactions, in which two small fragments are lost from the organic molecule. Of these reaction types, only the 1,2-eliminations will be discussed here, it being understood that examples of the mechanisms may be found, as appropriate, in other types of elimination reactions.

Concerted, bimolecular. Concerted bimolecular eliminations are characterized by second-order kinetics; they occur readily with powerful nucleophiles. A favoured stereochemical course (*trans*-elimination) involves a particular geometry, as shown, which requires that in the starting material the eliminated units be situated on opposite sides of the molecule. The olefinic product then must have the particular structure shown, rather than that of its geometric isomer. The relative extent to which the various bonds are formed and broken in the transition state varies considerably with the substrate.



starting material olefin product

Stepwise, bimolecular. If removal of the electrophilic fragment precedes the loss of the nucleophile, the reaction becomes stepwise and involves a carbanionic intermediate.

$$R_2CHCR_2Cl \underset{+BH,-B^-}{\overset{+B^-,-BH}{\rightleftarrows}} R_2C^- - CR_2Cl \xrightarrow{-Cl^-} R_2C = CR_2$$

starting material carbanion intermediate olefin product

Reaction by this path, which sometimes can be characterized by exchange of protons between the solvent and the starting material, is less stereospecific than the reaction by the concerted mechanism. This lessened stereospecificity is caused by the carbanion intermediate's not maintaining the rigid geometry characteristic of the concerted mechanism.

Carbonium ion intermediates

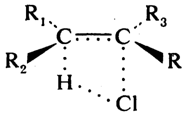
Stepwise, unimolecular. A carbonium ion produced by heterolysis (decomposition of a compound into oppositely charged particles or ions) may lose a proton, thereby effecting a 1,2-elimination reaction:

$$R_2CHCR_2Cl \underset{\substack{+Cl^- \\ slow}}{\overset{-Cl^-}{\rightleftarrows}} R_2CHCR_2{}^+ \xrightarrow{-H^+} R_2C{=}CR_2.$$

starting material carbonium ion olefin product

Such eliminations, which generally accompany nucleophilic substitutions, are promoted by electron-release to the carbonium-ion centre. The loss of the proton usually occurs in such a way as to give predominantly the thermodynamically more stable of the alternative products.

Cyclic. Some cyclic eliminations are fully concerted, but in others the loss of a nucleophilic or of an electrophilic component can be dominant. For example, the gas-phase pyrolysis (destructive heating) of alkyl halides shows the orientation and structure effects characteristic of unimolecular stepwise elimination reactions in solution. In such cases, the transition state (shown below), though still cyclic and preserving the stereochemistry, must involve greater stretching of the carbon–chlorine than of the carbon–hydrogen bond.



transition state

Expanding the valence shell

Nucleophilic replacements in complexes of metals. Stable compounds with more than four groups bonded to a central atom (the situation commonly encountered in compounds of carbon) are formed by elements in the second and higher rows of the periodic table of the elements. Mechanisms of reactions of these compounds therefore become more complex on stereochemical grounds alone. Furthermore, the energy levels of electron paths (orbitals) which can accommodate the bonding electrons of the reacting atom have become closer in these compounds, and reactions involving the formation of new bonds by expansion of the valency shell of this atom often become more readily accessible. For example, nucleophilic attack on carbon tetrachloride is slow, whereas that on silicon tetrachloride is fast, because in the former compound the attacked atom (carbon) has reached its maximum stable coordination number (indicative of the size of the valence shell), whereas in the latter, the central atom (silicon) has not, and its valency shell can be expanded simply by the attachment of a nucleophile. A similar difference in the mechanisms of reactions of metal complexes is found, depending on whether or not the metal atoms are free to engage in valence-shell expansion.

Unimolecular, in octahedral complexes. Octahedral complexes of metals of the first transition series (elements from scandium to copper) have reached their maximum stable coordination number, six. Accordingly, many of their replacement reactions are believed to occur by dissociation to give an intermediate having only five groups bonded to the reaction centre. Several different types of kinetic behaviour have been recognized. The initial stage may be a rate-determining dissociation of the cobalt complex shown below, in which methanol is the solvent, "en" is ethylene diamine ($H_2NCH_2CH_2NH_2$), and N^- can be any of a variety of nucleophiles, including bromide, thiocyanate, and nitrate ion.

$$[Co(en)_2Cl_2]^+ \xrightarrow[\text{slow}]{-Cl^-} [Co(en)_2Cl]^{2+1} \xrightarrow[\text{fast}]{+N^-} [Co(en)_2ClN]^+$$

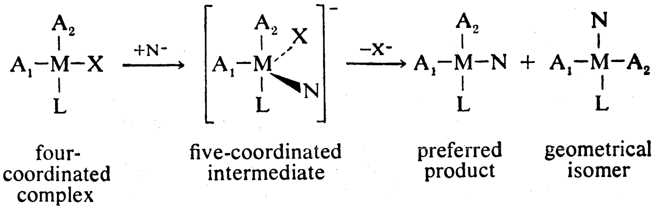
Alternatively the dissociative stage can be a pre-equilibrium, as in many replacements of water, as shown in the reaction below.

$$[Co(NH_3)_5(OH_2)]^{3+} \underset{+H_2O}{\overset{-H_2O}{\rightleftharpoons}} [Co(NH_3)_5]^{3+} \xrightarrow{+N^-} [Co(NH_3)_5N]^{3+}$$

The stereochemistry of these reaction paths is of great mechanistic significance, and varies both with the nature of the central metal atom and the nature of the attached groups (ligands).

Kinetic evidence

Bimolecular, in square planar complexes. Square planar four-coordinated complexes differ from their octahedral six-coordinated analogues in that they generally undergo bimolecular associative, rather than dissociative, nucleophilic displacements. Thus, for many reactions involving replacement of a ligand by a nucleophile in complexes of platinum(II) ions, a kinetic effect proportional to the concentration of the nucleophile can be identified, showing that the nucleophile is involved in the transition state. Furthermore, the stereochemical specificity of such reactions, as shown below, can be accommodated readily in terms of the five-coordinated associated intermediate, whereas a dissociative mechanism would be expected to result in the formation of a mixture of the two geometric isomers.



A ligand across from, or *trans* to, a replaceable group has a much greater influence on the rate of substitution than does the same substituent next to, or *cis* to, the replaceable group, and this *trans* effect helps to define the nature of the bonding in the transition state, because it suggests that only the *trans* substituent is in the same plane as the associated and departing group in the intermediate.

(P.B.D. de la M.)

CHEMICAL KINETICS

Chemical reactions are explained in terms of the atomic structure of matter and of the energy changes that can take place in atoms, bonding them into molecules or breaking up molecules to free or to rebond atoms into different molecules. Chemical reactions may be slow or fast, complicated or relatively simple, and the modes by which they proceed are the subject of chemical kinetics.

Preliminary considerations. In a broader view, molecules have a tendency to react with each other to produce other kinds of molecules on condition that the product molecules have less free energy than the reacting molecules had; *i.e.,* the products lie at a lower level of free energy than the reactants. An analogy is that of water always flowing downhill since, in the Earth's gravitational field, water at a higher level gives up energy as it falls to a lower level. Thus, all reactions for which the free energy, symbolized by ΔG, is negative have a tendency to occur. Accordingly, measurements of energy change suffice to indicate whether there will be a tendency for particular products to form when specific reactants are brought together under variable conditions. How fast a reaction will occur, however, depends on what sort of channels are available for the system to traverse in passing from the higher to the lower free energy state, and a study of the nature and effectiveness of these channels for reaction is best begun with a generalized symbolization of a reaction.

Rates of chemical reactions

In an actual reaction in which molecules of a substance A react with molecules of a substance B to form the molecules of substances C and D, usually the reverse reaction is also possible; that is, molecules of C and D react under suitable conditions to form molecules of A and B. If A, B, C, and D represent single molecules of those substances, a chemical equation may be written for the reaction: $A + B \leftrightharpoons C + D$, the half arrows indicating that the reaction is reversible. Of course, if molecules A and B are to react they must first collide, and certain other conditions must also obtain. The number of collisions will be proportional to the concentrations of A and B molecules; if the amount of A is doubled, twice as many encounters will take place. The same is true if the amount of B is doubled. Thus, a change in the concentrations of the substances will change the number of collisions, and the more numerous these collisions are per second, the faster the reaction will proceed (assuming that other determining factors, such as temperature, remain constant). Thus, the rate of the reaction, or its speed, can be calculated by simply multiplying the concentrations. If concentration is signified by bracketing the symbols for the molecules, $[A]$, $[B]$, $[C]$, and $[D]$, and if v_f symbolizes the forward rate, or the speed of reaction to the right in the equation, then this rate equals the product of the concentrations and of a proportionality factor, k_f, also called the specific reaction-rate constant: $v_f = k_f[A][B]$. Similarly, the backward rate is given by the equation $v_b = k_b[C][D]$. When the two rates are the same—*i.e.,* when $v_f = v_b$—the reaction, or, more correctly, the dynamic system of the two reactions, is said to be in equilibrium and $k_f[A][B] = k_b[C][D]$. Rearranging the equation, a ratio is obtained equal to a constant K, called the equilibrium constant; *i.e.,* the product of the concentrations of the products divided by the product of the concentrations of the reactants is equal to the forward reaction-rate constant divided by the backward reaction-rate constant:

$$\frac{[C][D]}{[A][B]} = \frac{k_f}{k_b} = K.$$

This is known as the reaction-equilibrium equation. The equilibrium constant K, in view of the energy changes necessary for a reaction to proceed, further relates to the free energy, ΔG, which in turn relates to the forward and backward free energy of activation. The energy of activation is a measure of the energy that the reacting molecules must have in order to react, and it is specific for each reaction. Molecules may collide forever without reacting, and all the familiar solids, liquids, and gases consist of molecules in intimate contact but not reacting. Activation energy for the burning of paper, for example, can be provided to the paper molecules by the flame of a match, itself stable until the heat from friction on the head provides enough energy to activate the chemical reaction of the ingredients. The activation energies, therefore, of the forward (ΔG_f) and backward (ΔG_b) reactions are related to the free energy by the equation $\Delta G = \Delta G_f - \Delta G_b$.

Early approaches to kinetics

Natural scientists have been impressed by the growth of crystals and of living things for centuries, but the first recorded, quantitative measurement of the rate of a chemical reaction was made in 1850, when it was noted that the plane of a beam of polarized light, passed through a solution of sucrose, was rotated by the solution and that the amount of rotation changed with time. This indicated that a chemical reaction was changing the sucrose into a new species of molecule (which rotated the polarized light to a different extent). The change in the concentration of sucrose can be measured and related to the passage of time to yield a specific rate constant for the reaction.

It was found that the rate of any reaction depends not simply on the concentration of the reactants but on these concentrations raised to various powers. This principle is known as the law of mass action. It was pointed out by the Norwegian chemists C.M. Guldberg and P. Waage around 1864–67 and precisely stated by a Dutch physical chemist, Jacobus Henricus van't Hoff, in 1877. In 1889 Svante Arrhenius, a Swedish chemist, published his conclusions that molecules must get into an active state before they can become reactive (*i.e.,* their energies must be raised to activation levels before they will react). Van't Hoff had already obtained the result for the equilibrium constant, *K,* by calculations involving the energy of reaction, the absolute temperature, and the gas constant (derived from energy relationships observed in the behaviour of gases). Accordingly, Arrhenius was able to formulate an equation (named after him) for the specific rate constant of a chemical reaction that included the energy of activation.

Arrhenius equation

Factors that affect the rate of reaction. Any change in the relative positions of atoms or molecules, such as the deformation of a solid, that takes place with an activation energy (*i.e.,* only after the energies of the reacting molecules have been increased to specific levels) is governed by the same general reaction-equilibrium equation (see above) as ordinary chemical reactions. Examples of solid reactions are the burning of rocket propellants and explosives by detonation, plastic deformation, phase transformations—such as freezing—and certain kinds of catalytic reactions. Many reactions that take place rapidly in the liquid state are sluggish or fail to take place at all in the gas phase.

Solvent. Conspicuous among these are changes that involve ionization (the breaking up of a substance into positively and negatively charged atoms or groups of atoms, called ions, and into electrons). Such reactions are greatly promoted by a solvent with high dielectric (electrically insulating) constant, such as water or dimethyl sulfoxide. Ions are also formed, however, in gases, such as the effluent gas from a burning rocket and as products of the oxidation of acetylene in the oxyacetylene torch. The temperatures involved are usually in the range from 2,000° to 2,500° C (3,600° to 4,500° F). Thus, certain reactions occurring at high temperatures can build up unstable ions at low concentrations in contrast to the heat-stable dissolved ions formed in high concentrations in ionizing solvents. Polar (*i.e.,* exhibiting positive and negative charges) surfaces, such as that of sodium chloride (common table salt), also promote the ionization of polar substances—such as chloroform, acetone, etc.—on their surface. In fact, many organic reactions occur by the formation of positive (called carbonium) ions in solution. Although the ions are present in very low concentration, careful electrical conductance measurements as well as a study of the products of reaction provide unmistakable evidence of ionization. When a metal is bound to an organic radical (a stable fragment of an organic molecule), there is a marked tendency for the molecule to decompose into a positive metal ion and a negative organic radical (called a carbanion) containing a negatively charged carbon.

Temperature. The effect of temperature on rates of reaction is important. Ordinarily, raising the temperature 10° (Celsius) approximately doubles the rate of most but not all reactions. This is because the chance that molecules are sufficiently energy rich to have achieved activation level and undergo reaction is greatly increased by an increase in temperature.

Pressure. Similarly, increasing the pressure on systems in which a reaction in the liquid or solid state is taking place usually changes the rate of reaction. Thus, if the reacting molecules in the activated state occupy more space than when they are not activated, an increase in the pressure will decrease the relative concentration of the activated state as compared with the nonactivated reactants and so slow down the rate of reaction. When the activated state is less voluminous, the rate of reaction will be increased by pressure. This is in accord with the Le Chatelier principle, which states that any system in equilibrium is shifted by a change in property, such as pressure or temperature, in such a direction as to reduce the stress of that change on the system. High pressure, for example, makes diamonds more stable to heat than graphite (diamond and graphite are the two crystalline forms of carbon) but, in spite of this, the transformation between graphite and diamond is slowed in either direction by pressure because the activated state of a carbon atom is more voluminous than a carbon atom in either diamond or graphite.

Le Chatelier principle

The effect of pressure on gaseous reactions is quite different. Here pressure always increases the concentration of reactants (*i.e.,* the number of molecules of gas per unit volume) and so increases the rate of all resulting reactions that involve more than one gaseous reactant because the frequency of forming the activated state is thereby increased. The pressure effect, which becomes important for condensed (solid and liquid) phases, is related to a difference between the actual volume of tightly packed reactants and the volume of the activated state. This effect is present for gases, too, but it plays an insignificant role in comparison to the role played by increase of concentration.

Catalysts. In catalytic reactions a minute amount of a substance (the catalyst) is added to the mixture of reactants to alter the rate of reaction enormously, usually increasing it, without being itself used up. If the catalyst is not soluble in the reaction mixture, the reaction takes place on the surface of the catalyst at activation levels far lower than are necessary without the catalyst. Thus, surface atoms are involved, and if one or more reactants is gaseous, the pressure effect will act to increase the concentration of reactants, as in a gas, and this will overshadow other pressure effects.

Collisions. Further development of the theory of reaction rates calls for elucidation of the Arrhenius equation. The collision theory, though incomplete, represented a notable advance for such reactions. For example, a hydrogen gas molecule, H_2, reacts with a gaseous iodine molecule, I_2, to give hydrogen iodide molecules, HI, according to the equation $H_2 + I_2 = 2HI$. It can be demonstrated that to exchange atoms in this reaction, the activated state is a four-atom complex in which the hydrogen molecule is sandwiched between two attacking iodine atoms formed by the dissociation of an iodine molecule. The activation energy is equal to the work expended in building the activated complex of atoms and consists of the energy required to dissociate an iodine molecule into atoms plus additional energy expended by the two iodine atoms in the breaking of the hydrogen-molecule bond. Since the activated state consists of the four atoms in collision exactly as it would if it were formed by the direct collision of a hydrogen molecule with an iodine molecule, it necessarily obeys the same collision formula; *i.e.,* a formula that calculates the number of collisions per second (here, between H_2 and I_2) and the chance that such a collision will be energetic enough to be successful. The collision formula is quite accurate for some reactions, but for others the predicted rate exceeds the experimental value by a factor of 10^8. This has led to the rewriting of the formula to include an empirical steric factor; *i.e.,* a factor related to the actual molecular structure.

Activated complexes

Early attempts to understand first-order reactions also ran into difficulties. It was supposed that individual molecules got into the activated state through absorbing radiant energy (called the radiation hypothesis), but this was soon found to be much too slow. It was then (1922) pointed out that activation by collisions alone would explain unimolecular reactions on condition that deactivation of the activated molecules, also by collision, was much faster than the rate at which they were used up in reaction. This point of view was developed (1928) into a useful theory.

In the reaction between hydrogen and chlorine molecules (H_2 and Cl_2) to form hydrogen chloride, HCl, the equation is $H_2 + Cl_2 \rightarrow 2HCl$, and it might be supposed that the same course would be followed as in the reaction between hydrogen and iodine, but this is not so. The preliminary dissociation into chlorine atoms, $Cl_2 \rightarrow 2Cl$, is followed by the reactions $Cl + H_2 \rightarrow HCl + H$ and $H + Cl_2 \rightarrow HCl + Cl$; these two processes constitute a

Chain reactions

chain reaction because a product of each is a starting material for the other. This chain is prevented from becoming infinitely long by chain-breaking reactions involving atom recombinations with another molecule or a surface, symbolized as M, such as $Cl + Cl + M \rightarrow Cl_2 + M$ and $H + H + M \rightarrow H_2 + M$. That the chain of reactions may be very rapid is shown by shining light on a hydrogen–chlorine mixture causing it to explode: the heat evolved as the chain reaction proceeds does not have time to diffuse away through the walls of the vessel before it raises the temperature of the reaction mixture to a point at which the resulting pressure bursts the container.

Factors that affect the order of the reaction. The degree to which the concentration of a reactant enters into the equation for the rate of reaction is called the order, with respect to this reactant. Thus, in the equation $H_2 + I_2 \rightarrow 2HI$, the reaction is of the first order with respect to each of the concentrations $[H_2]$, $[I_2]$.

In this example, the order of the reaction agrees with the stoichiometry. This is frequently not true, however; for example, the stoichiometry of the equations $H_2 + Cl_2 \rightarrow 2HCl$ and $Cl_2 \rightarrow 2Cl$ is the same, yet, because hydrogen and chlorine react by way of a chain of reactions, the reaction order is quite different. Thus, the equation $Cl + H_2 \rightarrow HCl + H$ is half order with respect to Cl_2 and first order with respect to H_2, while the equation $H + Cl_2 \rightarrow HCl + Cl$ is half order with respect to H_2 and first order with respect to Cl_2, so that the order of the overall chain process is complicated and dependent upon the conditions under which it occurs. In general, there is no way of deducing the order of reaction from the stoichiometry of the equation that expresses the overall process. Much of the effort devoted to the study of reaction rates goes into the establishing of the mechanism of the rate-determining step or steps. The overall order of reaction is further complicated if the surfaces catalyze the reaction.

Effects of surfaces

The degree to which the rate of reaction is influenced by a particular catalytic surface is determined by actually varying the ratio of reactant collisions with the surface to collisions with gaseous molecules. This is readily done for glass surfaces by packing the reactant vessel with glass wool; by varying the amount of glass wool it is easy to multiply manyfold the amount of surface with which reactants collide. If the packing produces no noticeable effect when the overall rate is large, it is safe to ignore the effect of the vessel's surface in the absence of packing and to interpret the rate as occurring entirely in the gas phase. Conversely, if the effect of packing is to speed up or slow down the reaction, it is possible to study the surface-catalyzed reaction.

The order of reaction under one set of conditions may be completely altered if the conditions are changed, even though there is no change in mechanism. If one reactant is in great excess over the others, for example, it will not seem to vary and may be left out of the rate equation altogether, even though the other concentrations may change by orders of magnitude. The reaction order is thus not to be confused with what is called the molecularity of the reaction: the molecularity is the number of molecules of each kind in the activated complex. For each elementary rate process the molecularity is fixed, and only a shift to a new mechanism can change it. When the molecularity of a substance in the activated state exceeds its molecularity in the stoichiometric equation for the reaction, that substance is said to be acting as a catalyst. Also, when a process occurs by a chain of reactions, the order with respect to a particular reactant may be fractional or even negative. Only when the effects of wide variations in the concentrations of the reactants and other conditions in a reaction have been studied is it possible to be confident that an observed order of reaction does in fact represent the molecularity of the activated state. As shown below, in more complicated cases, such as in chain reactions, the whole idea of associating a unique integral order with a rate of reaction ceases to be useful.

The equilibrium constant. A fundamental question that must be answered if chemical processes are to be controlled is not only how fast a system reaches equilibrium but under what circumstances equilibrium yields the particular product sought in recoverable amounts. This question is comparatively easy to answer for phase transformations. Thus, to obtain ice, water must be cooled below the freezing point. Even this, however, is not quite as simple as it sounds, since by increasing the pressure, the temperature at which ice melts increases. There are seven kinds of ice; *i.e.,* water freezes into seven different forms. It happens that ordinary ice is readily obtained by cooling, but if one of the other seven kinds is wanted, in general not only must the system be cooled but the pressure must be adjusted so that it falls in the range of stability of the ice that is sought. A similar problem occurs when diamonds are made from graphite. A diamond is unstable at ordinary temperatures and pressures, but the rate at which it changes into graphite is far too slow to cause worry. If the pressure is increased to 15,000 atmospheres (one atmosphere equals about 14.7 pounds per square inch), graphite becomes unstable and tends to change into diamond at room temperature. As before, however, the rate is much too slow to be interesting unless a catalyst is added or the temperature is greatly increased. The work, or the free energy, required to change graphite into diamond can be calculated according to a thermodynamic equation that relates measurements of temperature, pressure, and volume; at pressures and temperatures when the two phases are in equilibrium, the work required to change one phase into the other is zero. Thus, thermodynamics predicts how temperature and pressure can be manipulated to make a desired phase stable.

Thermodynamics and kinetics

A dramatic example of the consequences of studying an equilibrium is the case represented by the equation $3H_2 + N_2 \rightarrow 2NH_3$. Historically, this study has had awesome consequences because, had German scientists not discovered and used this reaction in the Haber process for making ammonia in 1914, Germany would have been dependent on Chile saltpetre for explosives, and a blockade by the Royal Navy might have made World War I impossible.

Using thermodynamics, the equilibrium constant can be calculated from the pressures of hydrogen, nitrogen, and ammonia, and the work required to change three moles of hydrogen and one mole of nitrogen into two moles of ammonia can be obtained as a function of pressure and temperature. Thus, the circumstances required to produce recoverable amounts of ammonia can be determined. In other words, thermodynamics indicates whether a reaction is possible and kinetics, the speed with which it will occur.

Complex reactions. Chemical reactions proceed by all physically possible mechanisms, and the measured rate is the sum of all the possible rates. Often, one mechanism under a given set of conditions is so much faster than all others as to allow the latter to be ignored, without appreciably affecting the accuracy of the calculations. It is in this sense that a preferred mechanism can be defined, and, theoretically, the preferred mechanism can be singled out only by estimating all rates and selecting the highest. If two velocities are equal at some temperature and much faster than all others, the one with the steepest temperature coefficient will be the important one at high temperatures and the other at low temperatures. In a case in which there are only two reactions in a sequence, and they become equal at some temperature, the one with the lowest temperature coefficient will dominate at low temperatures while the other dominates in the high-temperature range. This temperature dependence provides a ready way of distinguishing between two reactions acting sequentially and in parallel.

Competing reactions

Mass spectrography provides a beautiful example of competing reactions. When a molecule is struck by a fast electron or an energetic photon, it is frequently left as a positive ion in an energy-rich state. The mass spectrograph is designed so it records the final product ions about 10^{-6} second after the primary ionization. In the meantime, the parent ion is breaking into all of the various possible products at rates that can be calculated from changes in bond energies, and the percentage of each product ion formed can be calculated.

If a reaction intermediate is formed that can attack more than one species, the relative amounts of products are in proportion to the rates of reaction with the various species.

If a solid is thought of as a giant molecule, its action as a catalyst does not differ in principle from that of a homogeneous catalyst. When the atoms of a solid metal have their electron shells only partly filled, as in the transition elements, the crowded electrons leak into these partly empty shells during the period of intense electron congestion of the activated complex and, in the course of doing so, lower the activation energy speeding up the reaction.

On the other hand, certain ionic substances, such as clays, can act by breaking hydrocarbons into positive and negative ions, which then reassemble into molecules with new properties.

Enzymes are molecules that occur in living systems and that have structural groups that facilitate reactions, as do ordinary catalysts. One view of their behaviour is that they have the additional property of mechanically straining certain bonds lying between two attachments to the enzyme and giving the distorted bond an increased reactivity. This same strain, resulting from changes in the shape of the protein molecules, is promoted by environmental changes and is exhibited in muscles, in the receptors of taste and smell, and in various immune reactions. Study of these conformation changes reveals that they are volume changes because pressure changes slow down or speed up the process in question.

Chain reactions are another interesting aspect of chemical reactions, and it can be shown that there is no single simple order of reaction for a chain of reactions.

Detonations proceed by a quite different mechanism. A simple model will help explain how an energy-rich molecule may be made to yield its energy to reinforce a decaying shock wave, which is the essence of a detonation (see Figure 3).

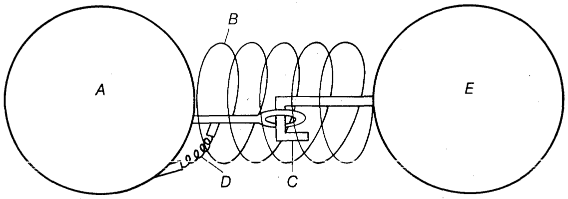


Figure 3: Model of molecular structure that explains detonation (see text).

The two parts of a molecule A and E may be thought of as being separated by a compressed spring B. If the amplitude of vibrations reaches a critical value, corresponding to the activation energy, so that during compression of B the spring D flips the catch C, then A and E will fly apart with the energy stored in the compressed spring B—*i.e.,* the heat of reaction.

Mechanism of detonations

If many such molecules are close packed to form a cylinder and a sharp blow is delivered to one end of the cylinder, a compression wave will pass along the cylinder axis. This shock wave will pass over molecules having random phases and random amplitudes, and some of them will absorb the extra energy that permits them to flip their catch—*i.e.,* to react. If the high pressure and high temperature of the shock wave remain, all of the molecules will react. These reacting molecules release their heat of reaction to enhance the energy in the shock wave. The energy released to the actual movement (called translational degrees of freedom of the molecules) determines the pressure and temperature of the shock wave, and that energy will just balance the energy flowing into the internal degrees of freedom during the period (called the steady state) when the shock wave is advancing at a fixed speed. Since sound is a wave of compression, the detonation will travel through the medium with the velocity of sound of molecules in the shock wave. If the diameter of the detonating cylinder is finite, an appreciable amount of the energy driving the shock wave leaks out through the side, decreasing the shock-wave velocity. The magnitude of this velocity decrease provides a method of measuring the length of the reaction zone.

Experimental study of chemical kinetics. *Measurement of reaction rates.* When reactions proceed at rates that permit making the measurements, many methods of study can be employed. The problem is one of measuring composition by some physical or chemical procedure. If the rate is fast compared with available methods of measurement, it may still be followed by interrupting the reaction by a sudden change—such as lowering the temperature, removal of the catalyst or one of the reactants, or by adding an inhibitor, a substance that drastically slows down a reaction, in effect, a catalyst with a negative effect. Common methods involve measurement of pressure changes, volume changes, and conductivity; more sophisticated methods include polarimetry; circular birefringence; refractometry, colorimetry, and spectrophotometry; potentiometry; polarography; and mass spectrometry. For a method to be useful the property being measured must change (*i.e.,* increase or decrease) with the extent of reaction, preferably in a linear or near-linear fashion—*i.e.,* in direct ratio.

Flowing systems

The study of a continuous flow of reactants through solid catalysts has long been common practice. Since the contact time can be varied, the method provides a way of measuring fast catalytic processes widely applicable in industry. In certain systems, the chemical reaction proceeds to completion within the time required for physical mixing of the reactants. One fruitful method of studying such a system is to displace it from equilibrium by sudden temperature or pressure increases, but some rates are too high or too low to be measured by any method.

Competition methods, used to measure fast reactions, are exactly what the term implies: the reaction is made to compete with a physical process having a known rate, or an equilibrium is displaced by a perturbation (disturbing effect), and the shift in equilibrium (equilibration) measured directly. In polarography, for example, this competition is between the formation of a reducible ion (an ion capable of acquiring electrons, thereby raising its negative charge) and its diffusion to the cathode. Most available information on electron transfer between transition-metal complexes has been found in this way. Proton magnetic resonance is illustrative of a method applicable to various nuclei having magnetic moments (the property comparable to the magnetism of a minute bar magnet). Measuring nuclear magnetic resonance is much like measuring the precession of a spinning top in a gravitational field.

Determining the order of reactions. A general picture of how reactions take place will help in attaching the proper importance to a determination of an order of reaction. The potential energy between any collection or grouping of atoms is a function of the smallest number of distances that fix the configuration of the group complex. If the potential energy is plotted as a function of the distances required to fix the configurations, a potential hypersurface in configuration space is obtained. Although this hypersurface ordinarily involves more than three dimensions, it will be convenient to use the language of three dimensions. Molecules represent valleys or basins on this surface, and a chemical reaction corresponds to passing from one low region to the other over the intermediate barriers. There are, in general, several passes that can be taken between low regions, and the traverse through a particular pass may involve crossing over a succession of passes separated by more or less deep depressions. These depressions are intermediate compounds. If one of these passes along a traverse is enough higher than any other along the traverse, and enough lower than for any other traverse, this particular pass will constitute the point of no return for this reaction, and the molecularity will also be the order of the reaction, provided all reactants are nearly enough at the same concentration so that their concentrations all change measurably during the reaction. If these conditions are not fulfilled, there is no unique activated complex and the order of reaction involves some kind of blend of rates for a variety of activated complexes, which can only accidentally be a measure of the molecularity of a particular activated complex. In chain reactions, the measured rate is a blend of a variety of rates, in general involving elementary processes from different potential surfaces. Accordingly, for such reactions, only by changing conditions drastically can a particular elementary process be made sufficiently prominent in the overall process to allow the corresponding specific rate constant to be de-

Reaction pathways

termined. When the overall rate for a chain reaction is broken into its elementary mechanisms, difficult though this may be, they are all found to involve the concentration of the separate reactants in the activated complex to an integral power.

Only for the simplest of reactions can one expect the measured order to reveal the molecularity of an elementary reaction. Nevertheless, by skillfully arranging conditions the specific rates of many elementary processes have been determined. For an elementary reaction, the reactants are in effective equilibrium with the activated complex. If equivalent amounts of reactants are chosen, the reactants will always remain in fixed proportion with respect to each other.

Bioluminescence

Elucidating reaction mechanisms. Even for very complex living systems, some observable property is controlled by a rate-determining step. It is often possible to gain real insight into what is happening by studying the effect of increasing the temperature and pressure and by changing the composition of the system. This is illustrated by studies made on bioluminescence. Luminescent bacteria give off almost no light at the freezing point; as the temperature is raised the luminescence goes through a maximum near 20° C (68° F), depending on the species, and again sinks to near zero at blood temperature. The luminescence is caused by the oxidation of luciferin by oxygen, in the presence of the enzyme luciferase, a reaction in which oxygen molecules accept electrons from the luciferin. In the process, two protons (hydrogen nuclei, or hydrogen ions with positive charge) from the reduced luciferin go into solution and the remaining two electrons are picked up by oxygen with the help of the enzyme luciferase. Nine times out of ten the oxygen picks off the less tightly bound electrons of the luciferin molecule, but once out of ten times it picks off a lower lying electron (one bonded more tightly into the structure), leaving one in a higher level, which then drops into the hole below, emitting a quantum of light. The intensity of the emitted light measures the overall rate of reaction, which in turn depends principally on the state of the enzyme. Thus, there is a ready means of studying the state of a representative oxidative enzyme *in vivo.*

Luminescence provides penetrating insight into many aspects of biology. The rise in luminescence with temperature is to be expected of a chemical reaction involving an activation energy. The maximum in luminescent intensity followed by a decline in luminescence arises from inactivation of the enzyme, causing a steep drop with temperature that more than overbalances the increase in the rate of the remaining enzyme that is still active. Because the enzyme in the activated complex is more voluminous than the free native enzyme, the increase it causes in hydrostatic pressure decreases the light intensity in the low temperature range lying below the luminescence maximum. Above the maximum, the hydrostatic pressure increases the light intensity because the inactive enzyme is more voluminous than it is in the activated complex.

Chemicals such as the alcohols, the ethers, urethane, and even the noble gas xenon under high pressure introduce hydrophobic (water-repelling) groups into cells and cause the enzymes to unfold and become inactive in a way that is parallel to anesthesia. On the other hand, the substances sulfanilamide and *p*-aminobenzoic acid inhibit luminescence by competing with the reactants for the prosthetic groups, and they are without influence on the unfolding of the enzyme. These results were all deduced from a quantitative fitting of equilibrium theory and activated complex theory to the data and are borne out by work on the purified luciferase and luciferin.

It is convenient to think of solids as giant molecules and of their plastic deformation as a molecular isomerization (*i.e.,* as the same components forced into a spatial rearrangement), to which reaction-rate theory applies just as it does to any other molecule. A stress on a solid, whether an electrical or mechanical one, always acts to lower the energy of activation for deformation. Only when the stresses are small is the relaxation (*i.e.,* the return of the deformed solid to its original structure) proportional to the stress, and two or more stresses then—and only then—obey reciprocal relations as defined by thermodynamics. Reaction-rate theory, however, is applicable for both large and small stresses. In fact, all deformations of matter involving activation energies can be studied and predicted using reaction-rate theory. (H.Ey.)

CHEMICAL RELAXATION PHENOMENA

The term relaxation is used by chemists and physicists to describe the interval or time lag between the application of an external stress to a system—that is, to an aggregation of matter—and its response. The relaxation effect may be caused by a redistribution of energy among the nuclear, electronic, vibrational, and rotational energy states of the atoms and molecules that comprise the system, or it may result from a shift in the ratio of the number of product molecules to the number of reactant molecules (those initially taking part) in a chemical reaction. The measurement of relaxation times can provide many insights into atomic and molecular structures and into the rates and mechanisms of chemical reactions.

Historical survey. The word relaxation was originally applied to a molecular process by the English physicist James Clerk Maxwell. In a paper "On the Dynamical Theory of Gases," which he presented in 1866, Maxwell referred to the time required for the elastic force produced when fluids are distorted to diminish or decay to $1/e$ (e is the base of the natural logarithm system) times its initial value as the "time of relaxation" of the elastic force. The earliest suggestion of a chemical relaxation effect is contained in a dissertation (Berlin, 1910) based on research directed by the German physical chemist Walther Nernst. Measurements of sound propagation through the gas nitrogen tetroxide, which breaks up, or dissociates, into nitrogen dioxide, led Nernst to suggest that experiments at frequencies at which the dissociation reaction could not keep pace with the temperature and pressure variations that occur within a sound wave, would permit evaluation of the dissociation rate. Ten years later, at a meeting of the Prussian Academy of Sciences, Albert Einstein presented a paper in which he described the various theoretical aspects of this relaxation effect.

Dielectric relaxation

The detection of the chemical relaxation effect predicted by Nernst and Einstein did not become technically feasible until the last half of the 20th century. In the first half of the century physicists and chemists in studying relaxation concentrated on physical relaxation processes. Peter Debye referred to the time required for dipolar molecules (ones whose charges are unevenly distributed) to orient themselves in an alternating electric field as dielectric relaxation. Sound absorption by gases was used to investigate energy transfer from translational, or displacement in space, to rotational (spinning and tumbling) and vibrational (oscillations within the molecule) degrees of freedom, the three independent forms of motion for a molecule. The former requires only a few molecular collisions, but the transfer of energy between translational and vibrational modes may require thousands of collisions. Because the processes are not instantaneous but time dependent, relaxation effects are observed. Their measurement provides information about molecular bonding and structure. Chemical relaxation was rediscovered by the German physical chemist Manfred Eigen in 1954. Since then, technological advances have permitted the development of techniques for the measurement of relaxation times covering the entire range of molecular processes and chemical reactivity.

The great variety of relaxation phenomena and of the techniques developed for their study precludes a comprehensive survey. To facilitate a general discussion, the relaxing system, its initial and final states, the nature of the disturbance, and the system's response are considered separately. Examples are cited that emphasize the important features of relaxation phenomena and illustrate the variety of information that can be obtained from their study. A moderately detailed description of one relaxation technique, the temperature-jump method, is used to summarize the discussion.

Relaxation mechanisms. The chemical relaxation of nitrogen tetroxide is easy to visualize, and it illustrates principles common to all relaxation phenomena. Nitrogen tetroxide (formula N_2O_4) actually is a dimer (a molecule

Dimers and monomers

formed from two similar constituents called monomers) that dissociates into two molecules of nitrogen dioxide (formula NO_2). The monomer and dimer are easily distinguishable: the former is a brown gas; the latter is a colourless gas. The product and reactants exist in equilibrium, represented by the reversible reaction,

$$\underset{\text{1 molecule}}{N_2O_4} \rightleftarrows \underset{\text{2 molecules}}{2NO_2}.$$

At ambient (room) temperature and atmospheric pressure, approximately 80 percent of the molecules in the mixture are dimers and the remaining molecules are monomers. The distribution of molecules between the two forms remains unchanged as long as the temperature and pressure are held constant. If the external conditions are altered, then the ratio of monomers to dimers will adjust to a new value. The dependence of the equilibrium on pressure is intuitively understandable as follows: to a good approximation the volume that a gas occupies at a given pressure and temperature depends directly on the number of gas molecules. The dissociation of one molecule of nitrogen tetroxide into two molecules of nitrogen dioxide entails an expansion of the gas, a doubling of molecules, which is opposed by the external pressure. If the external pressure is increased, the system acts to relieve the stress by reducing its volume; *i.e.,* by combining monomers to form dimers and thus reducing the number of molecules. The equilibrium shifts in favour of dimers under increased pressure, and in favour of monomers under reduced pressure. At any steady pressure, the ratio of the two forms eventually becomes constant.

Chemical relaxation results from the inability of systems at equilibria to respond instantaneously to changes in external conditions. The rate of reestablishment of equilibrium, or re-equilibration, is limited by the concentrations of the reactants and their reactivities. At any specified temperature and pressure, there is a definite probability per unit time that a nitrogen tetroxide molecule will dissociate into two nitrogen dioxide molecules and that the latter will recombine to form a dimer. The average lifetime of a nitrogen tetroxide molecule at ambient temperature and atmospheric pressure, for example, is about one-third of a microsecond (one-millionth of a second). The product of the reciprocal of the average lifetime times the concentration of nitrogen tetroxide molecules gives the rate at which they dissociate. At equilibrium there is no net change in the number of nitrogen tetroxide molecules, because their dissociation rate is exactly balanced by the rate at which they are being re-formed through association of nitrogen dioxide molecules. If the external conditions are altered, the reactivities of the monomer and dimer change instantaneously, but their concentrations change at a finite rate until the balance between the association and dissociation rates is reestablished.

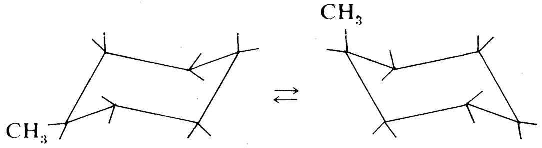
Sound waves in gases

Sound propagating through a gas can be pictured as a pressure wave whose alternating increase and falling off of pressure, called a sinusoidal variation of pressure, with time at any point in the medium is accompanied by a corresponding fluctuation in the temperature. The effect of the varying temperature and pressure of a sound wave moving through nitrogen tetroxide gas on the dissociation of nitrogen tetroxide depends on the frequency of that sound wave. When the pressure oscillates slowly enough, the dissociation reaction will remain at equilibrium with the oscillation; that is, the extremes in the monomer–dimer ratio will coincide with the extremes of pressure and temperature. If, on the other hand, the pressure fluctuates too rapidly for the reaction to follow, the ratio of monomers to dimers will remain constant at the equilibrium value for the ambient temperature and pressure; but at intermediate frequencies, a relaxation effect may be observed, and a readjustment of the chemical equilibrium will lag behind the pressure variation within the gas.

The relaxing chemical equilibrium results both in the absorption of sound by the gas and in dispersion of, or changes in, the sound velocity. Measurement of either of these effects permits evaluation of the relaxation time. The maximum absorption of sound occurs, for example, when the angular frequency (two pi times cycles per second) of the sound wave equals the reciprocal of the relaxation time. The relaxation time can then in turn be related to the mechanism of the chemical reaction and to the reactivities of the reactants.

The relaxing system. Relaxation may occur between any two allowed energy states of nuclei, atoms, or molecules in the solid, liquid, or gas phase. A distinction has already been made between chemical relaxation, which involves a transformation between two chemically distinguishable molecules such as the dissociation of nitrogen tetroxide, and physical processes such as the transfer of energy between translational and vibrational states of a molecule displayed by sound absorption in a homogeneous gas. Although it is useful to classify relaxation processes as chemical or molecular, the distinction between them depends on the height of the energy barrier separating the chemical species, and it becomes blurred when structural isomerizations are considered. Liquid methyl-cyclohexane, for example, absorbs sound of ultrahigh frequency. The relaxation effect is attributed to an isomerization (change in structure) between two forms of the molecule called the axial and equatorial chair forms, as shown below:

Isomerizations



equatorial form of methylcyclohexane

axial form of methylcyclohexane

In the axial form, the methyl group lies perpendicular to the principal axis of the carbon ring, whereas in the equatorial form the methyl group lies in the plane of the ring. Whether or not the interconversion is considered a chemical or a molecular relaxation process is largely a matter of definition.

Atomic nuclei may exhibit relaxation effects. Some nuclei spin mechanically. Because nuclei are charged, there is a magnetic field associated with a spinning nucleus: it behaves like a simple bar magnet with a north and south pole. The nucleus is said to have a magnetic moment that will experience a force when placed in an external magnetic field. A hydrogen nucleus in an external magnetic field, for example, may orient its nuclear magnetic moment either parallel or antiparallel to the external field. The latter is a higher energy orientation, called the upper spin state. The equilibrium distribution of many hydrogen nuclei between the two spin states (parallel and antiparallel) can be perturbed (*i.e.,* changed) by the absorption of electromagnetic radiation of appropriate frequency. The system will then relax to the equilibrium distribution by time-dependent radiationless transitions of the hydrogen nuclei from the upper to the lower spin state. This process of returning to the equilibrium distribution is called spin-lattice relaxation, because the excess energy of the upper spin state is transferred to molecules surrounding the relaxing hydrogen nuclei as increased translational, rotational, or vibrational energy.

As with nuclei, atoms and molecules can be excited to higher energy states by the absorption of electromagnetic radiation. A nonequilibrium distribution of atoms or molecules in excited states is frequently accomplished by a technique, called flash photolysis, in which the system of atoms or molecules is subjected to an intense flash of visible or ultraviolet light. The excited species may undergo many fates, but if they decay to the equilibrium distribution between the ground, or lowest, states and the excited states of the original atoms or molecules, the system is said to have relaxed.

The word relaxation is sometimes used to describe the radiation of energy by individual molecules, atoms, or nuclei, rather than by a large number. A hydrogen nucleus, for example, may decay from the upper to the lower spin state by transferring radiant energy to a nearby hydrogen nucleus in the lower spin state. This exchange of spins is called spin-spin relaxation. It shortens the lifetime of an individual excited nucleus, but it does not restore the equilibrium distribution of parallel and antiparallel spins. Although it is convenient to think of an individual excited nucleus as relaxing, only the response of an excited population of many nuclei can be measured. This usage of the term relaxation obscures the most useful experimental feature of relaxation processes.

Initial and final states. In virtually all relaxation experiments a thermodynamic equilibrium state is disturbed, and the time required for re-equilibration is measured. The practical advantage of starting with a system at equilibrium is most apparent in the study of chemical reactions in solution. Nearly all of the elementary steps in chemical reactions, such as transfers of protons and electrons from one molecule to another, occur in less than a millisecond, and yet as late as the 1960s solution reactions with half-times (time in which the reaction is half completed) shorter than a millisecond could not be studied. This limit was imposed by the hydrodynamic problem of mixing two solutions. Reaction rates had been studied by mixing the reactants and monitoring the rate at which products appeared. The most elaborate mechanical mixing devices that have been built so far require a millisecond to initiate a solution reaction. Manfred Eigen was the first person to clearly perceive that mixing could be avoided by perturbing an equilibrium and watching it relax. His enormous contribution to the study of fast chemical reactions was recognized by the award of a Nobel Prize in 1967.

Perturbation

Instead of disturbing an equilibrium system, a stationary state may be perturbed. Many enzyme-catalyzed reactions, for example, are experimentally irreversible. Nevertheless, for much of the time course of the reaction, the chemical intermediates are present in a stationary state; that is, their concentrations do not change. The stationary state can be disturbed and the rate of its reestablishment may be used to deduce the lifetimes of the chemical intermediates. Combined rapid mixing and relaxation techniques have been used successfully in a study of catalysis by the enzyme ribonuclease.

Creation of the disturbance. Eigen has divided the methods used to disturb systems into indirect, or competition, methods and direct, or perturbation, methods. In the first method, the relaxing system is continuously disturbed. The competition between the disturbance and the relaxation process results in the establishment of a stationary state, from which information about the relaxation process must be inferred. Ultrasonic absorption is an example of a competition method. The competition between the temperature and pressure variations in the sound wave and the dissociation of nitrogen tetroxide sets up a stationary state in which re-equilibration of the chemical reaction lags behind the pressure fluctuations in the sound wave. The reactivities of the monomer and dimer are derived indirectly from measurements of sound absorption. Flash photolysis is an example of the second method, in which the system is momentarily perturbed. The molecules are electronically excited from the ground, or lowest and normal, energy state to higher energy states by the flash. Their return, or decay, to the ground state can be followed directly by monitoring the reemission of the absorbed light.

A chemical equilibrium can be disturbed by changing the pressure or temperature or by applying an electric field. If a volume change accompanies a chemical reaction, the ratio of products to reactants at equilibrium will depend on the pressure. The point at which equilibrium sets in will depend on temperature, if heat is absorbed or released in the reaction. The ratio will also depend on electric field strength, if the polarizabilities (change in orientation or position of electric charges) of the reactants and products are different. Nuclear and electronic states can be excited by the absorption of electromagnetic radiation, and the latter can also be excited thermally. Perturbation forces, when expressed mathematically in terms of strength and time, are called forcing functions. In principle, a forcing function may assume any form, but in practice, it must be easy to generate experimentally and to analyze mathematically. Examples of forcing functions are the sinusoidal temperature and pressure variations in a sound wave (charting the variations produces a curve called a sine curve—which varies from positive to negative values) and sinusoidally alternating electric fields, which are used in dielectric relaxation measurements. Other convenient forcing functions are step, or incremental, perturbations and rectangular pulses (pulses of which the strength rises nearly instantaneously, remains at the higher value for a period of time, and then rapidly returns to its initial value).

Step perturbations of the temperature and pressure can be produced in shock tubes. A gas at high pressure is separated by a membrane from the gas being studied at low pressure. When the membrane is burst, a plane pressure wave caused by the high-pressure driving gas moves through the low-pressure gas under study. Temperature increases of several thousand degrees may accompany moderate pressure shocks. The shock front travels through the gas with a velocity comparable to the mean molecular velocity, so that the width of the shock front is only a few mean free paths (average distances travelled by the molecules between collisions). As the shock passes, the translational energy of the molecules in the shock front is increased. The system relaxes as the excess energy is distributed by collisions to rotational and vibrational degrees of freedom.

Rectangular temperature perturbations (plotted on a graph these show up as a curve that periodically rises suddenly, stays constant for an interval, then drops suddenly to original value) can be produced in aqueous solutions of reacting systems by using microwaves to heat the solution. Water molecules can absorb energy of rotation at 10^{10} hertz (cycles per second). By concentrating the microwave energy in a small volume, an increase of several degrees in temperature can be obtained in one microsecond using pulses of radar. Since the radar generator can be repeatedly pulsed, coupling it with a continuous flow system improves the experimental accuracy by averaging over the period of the experiment.

Detection

Response of the system. Any of the techniques for disturbing an equilibrium can be combined with a variety of detection systems. Depending on the nature of the relaxation effect, it can be monitored by absorption or emission spectroscopy, by fluorometry, or by polarimetry. Conductance changes can be measured. Crystals are used to detect ultrasonic waves and to measure absorption effects.

While a priori there is no restriction on the magnitude of the displacement from equilibrium, in practice small disturbances are used to permit the application of a linear rate equation (terms denoting changes with time are to the first power). The rate of disappearance, for instance, of a small displacement from equilibrium is approximately proportional to the magnitude of the displacement. This relationship is given by the differential equation

$$-\frac{d}{dt}=(\Delta X)=\frac{\Delta X}{\tau}\cdot$$

Here, the displacement (ΔX) is the difference between the instantaneous and the equilibrium values of the relaxing property, which might be the kinetic energy of molecules behind a shock front or the concentration of a chemical reactant. The reciprocal of the constant of proportionality has units of time and is called the relaxation time (τ, *tau*). Since the equilibrium values may be time dependent, the solution of the rate equation depends on the form of the forcing function. Propagation of a sound wave through nitrogen tetroxide gas, for instance, causes a sinusoidal variation of the equilibrium concentrations of monomers and dimers with time. A great advantage of relaxation methods is that the response to small disturbances can be approximated by a first-order differential equation.

The relaxation time for a chemical process can be related to the reactivities of the reactants if the reaction mechanism is known. Conversely, it may be possible to deduce the reaction mechanism from the dependence of the relaxation time on reactant concentrations. If several chemical reactions are coupled, or if more than one vibrational state is excited, a spectrum of relaxation times may be observed. The relaxation times for the individual relaxation processes can be determined from the measured relaxation times, which are the normal modes for the coupled system.

Temperature-jump experiment. To summarize and to clarify this discussion, a temperature-jump relaxation experiment will be described. The name temperature jump is reserved for the relaxation technique in which a stepwise temperature perturbation is achieved by passing a large electric current through the solution under study and thus heating it almost instantaneously. Instrumentally, it is one of the simplest relaxation techniques. It is also the most generally useful method for the study of fast chemical reactions in solution.

Temperature-jump instrument

A typical temperature-jump instrument produces a temperature rise of approximately 8° C within five microseconds. The principles of this instrument are briefly explained as follows. A 0.05-microfarad capacitor is charged to between 30 and 40 kilovolts. The electrical energy stored on the capacitor is proportional to its capacitance and to the voltage squared. It is discharged through the reaction cell at time zero by closing a variable spark gap. The time required for dissipation of roughly 80 percent of the stored energy is given by the product of the capacitance times the cell resistance. The energy is dissipated through collisions between the ions, which conduct the discharge current through the solution and the solvent molecules. The rapid temperature increase causes a shift in the concentrations of reactive molecules in the solution to new equilibrium values. If this shift is accompanied by a colour change, the reaction rate can be monitored spectrophotometrically (*i.e.,* the change in the intensity of light of a selected wavelength with time is measured). The results are recorded on a storage oscilloscope for later display. Provided that the rise time of the temperature pulse is much shorter, and that the thermal re-equilibration time is much longer, than the response time of the chemical reaction being studied, the temperature jump can be approximated as a step perturbation. At times greater than zero, the equilibrium concentrations of the reactants remain constant at the values corresponding to the higher temperature. Consequently, the differential equation for the disappearance of the displacement of reactant X from equilibrium can be integrated to show that this value decays exponentially.

In the introduction to an article on the *Molecular Basis of Visual Excitation* the Nobel laureate George Wald wrote,

> I have often had cause to feel that my hands are cleverer than my head. That is a crude way of characterizing the dialectics of experimentation. When it is going well, it is like a quiet conversation with Nature. One asks a question and gets an answer; then one asks the next question, and gets the next answer. An experiment is a device to make Nature speak intelligibly. After that one has only to listen.

Relaxation phenomena afford a unique method for making nature speak intelligibly about rapid energy transfers and chemical reactions. They have only begun to be exploited, especially to probe the elementary steps in complex biochemical reactions. (L.D.F.)

CATALYSIS

The rates of chemical reactions—that is, the velocities at which they occur—depend upon a number of factors, including the chemical nature of the reacting species and the external conditions to which they are exposed. A particular phenomenon associated with the rates of chemical reactions that is of great theoretical and practical interest is catalysis. As noted earlier, this phenomenon is the acceleration of chemical reactions by substances not consumed in the reactions themselves—substances known as catalysts. The study of catalysis is of interest theoretically because of what it reveals about the fundamental nature of chemical reactions; in practice, the study of catalysis is important because many industrial processes depend upon catalysts for their success. Finally, the peculiar phenomenon of life would hardly be possible without the biological catalysts termed enzymes.

History. The term catalysis (from the Greek *kata-*, "down," and *lyein,* "loosen") was first employed by the great Swedish chemist Jöns Jacob Berzelius in 1835 to correlate a group of observations made by other chemists in the late 18th and early 19th centuries. These included: the enhanced conversion of starch to sugar by acids first observed by Gottlieb Sigismund Constantin Kirchhoff; Sir Humphry Davy's observations that platinum hastens the combustion of a variety of gases; the discovery of the stability of hydrogen peroxide in acid solution but its decomposition in the presence of alkali and such metals as manganese, silver, platinum, and gold; and the observation that the oxidation of alcohol to acetic acid is accomplished in the presence of finely divided platinum. The agents promoting these various reactions were termed catalysts, and Berzelius postulated a special, unknown catalytic force operating in such processes.

In 1834 the English scientist Michael Faraday had examined the power of a platinum plate to accomplish the recombination of gaseous hydrogen and oxygen, the products of electrolysis of water, and the retardation of that recombination by the presence of other gases, such as ethylene and carbon monoxide. Faraday maintained that essential for activity was a perfectly clean metallic surface (at which the retarding gases could compete with the reacting gases and so suppress activity), a concept that would later be shown to be generally important in catalysis.

Early applications of catalysis

Many of the primitive technical arts involved unconscious applications of catalysis. The fermentation of wine to acetic acid, the manufacture of soap from fats and alkalies, and the formation of ether from alcohol and sulfuric acid—all catalytic reactions—were well known in man's early history. Sulfuric acid prepared by firing mixtures of sulfur and nitre (sodium nitrate) was an early forerunner of the lead chamber process of sulfuric acid manufacture, in which sulfur dioxide oxidation was accelerated by the addition of oxides of nitrogen. (A mechanism for the latter process was suggested by Sir Humphry Davy in 1812 on the basis of experiments carried out by others.)

In 1850 the concept of a velocity of reaction was developed during studies of hydrolysis, or inversion, of cane sugar. The term inversion refers to the change in rotation undergone by monochromatic light when it is passed through the reaction system, a parameter that is easily measured, facilitating study of the reaction. It was found that the rate of inversion was, at any moment, proportional to the amount of cane sugar undergoing transformation and that the rate was accelerated by the presence of acids. (Later it was shown that the rate of inversion was directly proportional to the strength of the acid.) This work was, in part, the precursor of later studies of reaction velocity and the accelerating influence of higher temperature on that velocity by J.H. van't Hoff, Svante Arrhenius, and Wilhelm Ostwald, all of whom played leading roles in the developing science of physical chemistry. Ostwald's work on reaction velocities led him in the 1890s to define catalysts as substances that change the velocity of a given chemical reaction without modification of the energy factors of the reaction.

This statement of Ostwald was a memorable advance since it implied that catalysts do not change the position of equilibrium in a reaction. In 1877 Georges Lemoine had shown that the decomposition of hydriodic acid to hydrogen and iodine reached the same equilibrium point at 350° C, 19 percent, whether the reaction was carried out rapidly in the presence of platinum sponge or slowly in the gas phase. This observation has an important consequence: a catalyst for the forward process in a reaction is also a catalyst for the reverse reaction. P-E-M. Berthelot, the distinguished French chemist, had confirmed this observation in 1879 with liquid systems when he found that the reaction of organic acids and alcohols, called esterification, is catalyzed by the presence of small amounts of a strong inorganic acid, just as is the reverse process—the hydrolysis of esters (the reaction between an ester and water).

Industrial applications

The application of catalysts to industrial processes was undertaken deliberately in the 19th century. P. Phillips, an English chemist, patented the use of platinum to oxidize sulfur dioxide to sulfur trioxide with air. His process was employed for a time but was abandoned due to loss of activity by the platinum catalyst. Subsequently poisons in the reactants were found to be responsible, and the process became a technical success at the turn of the

20th century. In 1871 an industrial process was developed for the oxidation of hydrochloric acid to chlorine in the presence of cupric salts impregnated in clay brick. The chlorine obtained was employed in the manufacture of bleaching powder (a dry substance that releases chlorine on treatment with acid) by reaction with lime. Again, in this reaction, it was observed that the same equilibrium was reached in both directions. Furthermore, it was found that the lower the temperature, the greater the equilibrium content of chlorine; a working temperature of 450° C produced the maximum amount of chlorine in a convenient time. Toward the close of the 19th century, the classical studies of the eminent French chemist Paul Sabatier on the interaction of hydrogen with a wide variety of organic compounds were carried out using various metal catalysts; this research led to the development of a German patent for the hydrogenation of liquid unsaturated fats to solid saturated fats with nickel catalysts. The development of three important German catalytic processes had great impact on industry at the end of the 19th century and in the early decades of the 20th. One was the so-called contact process for producing sulfuric acid catalytically from the sulfur dioxide produced by smelting operations. Another was the catalytic method for the synthetic production of the valuable dyestuff indigo. The third was the catalytic combination of nitrogen and hydrogen for the production of ammonia—the Haber process for nitrogen fixation—developed by the chemist Fritz Haber.

Homogeneous catalysis. When the catalyst and the reacting substances are present together in a single state of matter, usually as a gas or a liquid, it is customary to classify these as cases of homogeneous catalysis. Oxides of nitrogen serve as catalysts for the oxidation of sulfur dioxide in the lead-chamber process for producing sulfuric acid, an instance of homogeneous catalysis in which the catalyst and reactants are gases. Traces of water vapour catalyze some gas reactions; for example, the interaction of carbon monoxide and oxygen, which proceeds only slowly in dry conditions. Sulfuric acid used as a catalyst for the formation of diethyl ether from ethyl alcohol is an example of homogeneous catalysis in the liquid phase (when the products, water and ether, are continuously removed by distillation); by this method, considerable quantities of alcohol can be converted to ether with a single charge of sulfuric acid. The inversion of cane sugar and the hydrolysis of esters by acid solutions (see above) also are examples of homogeneous catalysis in the liquid phase.

The oxidation of sodium sulfite solutions by dissolved oxygen is greatly accelerated by minute traces of copper ions in the homogeneous liquid system. This system is of special interest since it has been shown that the process is a chain reaction. In this case many thousands of molecules of sodium sulfite can be oxidized to sulfate if the initial activation process is produced by absorption of a limited number of quanta (discrete energy measures) of light. The best example of a light-initiated chain reaction is the photo-combination of hydrogen (H_2) and chlorine (Cl_2); as many as 1,000,000 molecules of hydrogen chloride can be formed by absorption of a single light quantum (designated $h\nu$). Here the sequence of reaction is

$$Cl_2 + h\nu \rightarrow 2Cl\cdot \tag{1}$$

$$Cl\cdot + H_2 \rightarrow HCl + H\cdot \tag{2}$$

$$H\cdot + Cl_2 \rightarrow HCl + Cl\cdot \tag{3}$$

with reactions (2) and (3) repeated over and over again. It is of interest to note that such chain reactions can be retarded by the presence of negative catalysts, more commonly termed inhibitors. These are materials that slow down the overall reaction by shortening the reaction chains, generally by entering into a non-chain reaction with one of the chemical components that maintain the chain. A wide variety of substances, including alcohols, sugars, and phenols, has been found to act as inhibitors of the oxidation of sulfite solutions.

A generalized treatment of homogeneous catalysis by acids and bases was given by the Danish physical chemist J.N. Brønsted in the mid-1920s on the basis of his concept of acids and bases. According to Brønsted, an acid is a molecule that can furnish a proton, whereas a base is a molecule that takes up a proton. On this assumption, the range of acids includes such varied materials as bisulfate ion, HSO_4^-; acetic acid, CH_3COOH; water, H_2O; hydronium ion, H_3O^+; ammonium ion, NH_4^+; and the corresponding bases are sulfate ion, SO_4^{2-}; acetate ion, CH_3COO^-; hydroxide ion, OH^-; water, H_2O; and ammonia, NH_3 (these substances accept protons to yield the listed acids). Brønsted studied a number of acid- and base-catalyzed reactions, including: (1) the acid-catalyzed hydrolysis of an ester, ethyl orthoacetate; (2) the basic catalysis of nitramide decomposition ($H_2N\text{–}NO_2 \rightarrow H_2O + N_2O$); and (3) the acid-base catalysis of the conversion (mutarotation) of glucose to a closely related form. In each case he observed a direct relationship between the velocity of the catalyzed reaction and the concentration of the catalyzing substance.

Basic mechanisms of homogeneous catalysis

Based in part on the ideas of Brønsted, a general scheme for a change of a substance A to another substance B catalyzed by a material C can be formulated thus:

$$A + C \underset{k_2}{\overset{k_1}{\rightleftharpoons}} Z \tag{1}$$

$$Z \xrightarrow{k_3} B + C \tag{2}$$

The designation Z refers to an intermediate stage, which is formed with a velocity indicated by k_1; Z can disappear either to reform $A + C$, with a velocity k_2, or it can decompose by path (2), with velocity k_3, to give the product B and regenerate the catalyst C. If k_3 is much greater than k_2 the intermediate Z is used up almost as quickly as it is formed. The product will then be formed at a rate governed by the expression $k_1[A][C]$, in which the square brackets indicate concentrations of reactant and catalyst. If k_2 is much larger than k_3, however, the velocity-determining process is the decomposition of Z, the rate of formation of the product being represented by the expression $k_3[Z]$, in which $[Z]$ is the concentration of the intermediate. Examples of both types of change have been studied.

In certain instances two or more catalysts present at the same time produce effects greater than either would produce alone. It is then customary to speak of promoter action. Thus, iron ions in solution fortify the action of copper ions in catalyzing a reaction between hydrogen peroxide and iodine. It is assumed that each catalyst activates only one of the reactants.

Homogeneous catalysis of industrial importance

The most important modern examples of homogeneous catalyses are found in the petrochemical industry. The oxo reaction is one such process; in this process carbon monoxide and hydrogen are added to olefins (unsaturated hydrocarbons) at around 150° C and 200 atmospheres of pressure to form aldehydes and alcohols, oxygen-containing organic compounds. A cobalt carbonyl catalyst $Co_2(CO)_8$ is employed; this hydrocarbon-soluble catalyst is believed to activate hydrogen by formation of $HCo(CO)_4$, which then reacts with the olefin. This reaction has led to a number of studies of organometallic chemistry. Copper, silver, and mercury cations (positively charged ions) and permanganate anions (negative ions) also are known to act as homogeneous catalysts for hydrogen activation. Palladium chloride is employed industrially in the catalytic oxidation of ethylene to acetaldehyde in the presence of cupric chloride. The palladium is presumed to be repeatedly converted from the salt to the free metal, the function of the cupric chloride being to participate in the re-formation of the palladium salt from the metal.

Phosphoric, sulfuric, sulfonic, and hydrobromic acids are important agents in the industrial processes of isomerization, polymerization, hydration, and dehydration, as well as in the classical esterification reactions. Free radicals (molecular fragments bearing unpaired electrons) that are generated by the decomposition of peroxides or metal alkyls also initiate homogeneous catalytic processes.

Heterogeneous catalysis. Many catalytic processes are known in which the catalyst and the reactants are not present in the same phase—that is, state of matter. These are known as heterogeneous catalytic reactions. They include reactions between gases or liquids or both at the surface of a solid catalyst. Since the surface is the place at which the reaction occurs, it generally is prepared in ways that produce large surface areas per unit of catalyst; finely divided metals, metal gauzes, metals incorporated into supporting matrices, and metallic films have all been used in modern heterogeneous catalysis. The metals themselves are used or they are converted to oxides, sulfides, or halides.

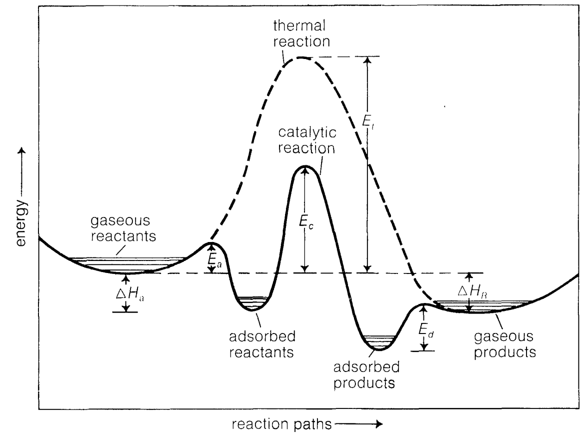


Figure 4: Energy profiles for catalytic and thermal (noncatalytic) reactions in the gaseous phase: E_c is activation energy for catalytic reaction; E_t is activation energy for thermal reaction; E_a is activation energy of adsorption of gaseous reactants; E_d is activation energy of desorption of gaseous products; ΔH_a is heat of chemisorption of reactants; ΔH_R is heat of overall reaction.

With solid catalysts, the reactants are chemisorbed (a portmanteau term for chemically adsorbed) by the catalyst. A catalyst is chosen that releases the products formed as readily as possible; otherwise the products remain on the catalyst surface and act as poisons to the process. Chemisorption can occur over a wide temperature range, the most effective temperature for adsorption depending on the nature of the catalyst. Thus, hydrogen is chemisorbed readily by many metals even at liquid air temperatures (below −180° C). With a series of hydrogenation–dehydrogenation catalysts—*e.g.,* zinc oxide–chromic oxide ($ZnO–Cr_2O_3$)—chemisorption of hydrogen often occurs above room temperature. Nitrogen is chemisorbed on synthetic ammonia-iron catalyst rapidly in the region above 400° C. It has been shown that iron films chemisorb nitrogen even at liquid air temperatures, with additional chemisorption found above room temperatures. It follows from such considerations that whereas physical adsorptions, which parallel the ease of liquefaction of the adsorbed substance, occur spontaneously, chemisorption, which involves the making and breaking of chemical bonds, often requires activation energies (energy needed to initiate reactions) as do uncatalyzed chemical processes. To be efficient catalytically, a process must involve energies of activation for all the steps involved that, at their maxima, are less than those required for the uncatalyzed reaction. This situation is illustrated graphically in Figure 4, for a hypothetical reaction, which can occur by either an uncatalyzed or a catalyzed route. Two competing proposals have been made concerning the mechanism of catalytic reactions at surfaces, and it has not been possible to choose between them. Originally, Irving Langmuir, the American physical chemist, proposed chemisorption of both reacting species at the surface, followed by interaction between adjacent species and evaporation of the products. An alternative proposal involves interaction between an impinging molecule and species already adsorbed on the surface. Subsequent developments have suggested various modes of attachment of the adsorbed and adsorbing species.

Basic mechanisms of heterogeneous catalysis

A major advance in the science of surface catalysis was the development of a method for determining the surface area of catalysts (and other materials) by measuring the multimolecular adsorption of nitrogen at liquid nitrogen temperatures or the adsorption of other gases close to their boiling points. It then became possible to calculate a quantity, designated V_m, which represents the volume of gas necessary to form a monolayer on the accessible surface; furthermore, the area of the surface can be determined from the known dimensions of the adsorbed molecules. It has also been found possible to titrate (measure quantitatively) the area of surfaces by chemisorption of gases.

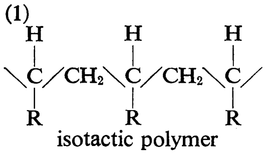
When measurements of surface areas became possible, it was seen at once that many constituents present in minor quantities in the main catalyst material—and known as promoters—could act by extending the effective surface area of the catalyst. It also was shown, however, that a promoter might produce an increase in the quality of the surface for the given reaction. Acting in a reverse direction are minor constituents of the reacting system or unwanted products of the reaction, which by preferential adsorption on the reaction sites and resistance to removal give rise to poisons for the process.

Much can be learned about mechanisms of surface processes by studying the behaviour of isotopic species of the reactants and products on the catalyst. An example of such use concerns the technically important synthesis of ammonia from its elements, the well-known Haber process on promoted-iron catalysts. The synthesis of ammonia involves three types of bonds, hydrogen–hydrogen, nitrogen–hydrogen, and nitrogen–nitrogen, all of which can be studied using isotopes of hydrogen and nitrogen. The first of these can be examined on the reduced-iron catalyst by following the progress of the reaction $H_2 + D_2 \rightarrow 2HD$ (in which D equals the deuterium atom, an isotope of hydrogen) on the surface. The reaction is found to occur rapidly even at liquid air temperatures. Nitrogen–hydrogen bond activation can be studied by following the reaction $NH_3 + ND_3 \rightarrow NH_2D + NHD_2$. This proceeds steadily at room temperatures. The reaction involving only N–N bonds, however, studied by following the process $^{14}N_2 + {}^{15}N_2 \rightarrow 2{}^{14}N–{}^{15}N$ (in which ^{14}N and ^{15}N are stable isotopes of nitrogen), is shown to proceed only at the higher temperatures of ammonia synthesis, around 400° C. From these data one concludes that the activation of the nitrogen molecules is the slow step (the process that limits the overall reaction) in ammonia synthesis. This conclusion is confirmed by measurement of rate of adsorption of nitrogen on the iron catalysts. Other, similar isotopic studies have yielded valuable information on the reactions of hydrocarbons, using deuterium and carbon-14 as the isotopic tracers.

Catalysis in stereoregular polymerization

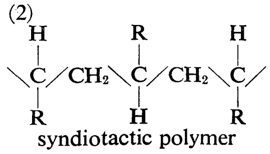
The importance of the concept of adsorption of reactants on the surface of catalysts has been greatly increased by the development of stereoregular polymerization processes—that is, methods that yield polymers whose molecules have definite three-dimensional patterns. Such processes were developed independently by the German chemist Karl Ziegler and the Italian Giulio Natta. An example is the polymerization of propylene with a titanium trichloride–alkyl aluminum catalyst. In the case of a generalized ethylenic compound, $CH_2 = CHR$, stereoregular polymerization may yield three different arrangements of the polymer; an isotactic polymer, a syndiotactic polymer, and an atactic polymer. These have the following arrangements of their molecular chains:

(1)



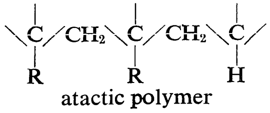
isotactic polymer

(2)



syndiotactic polymer

(3)



atactic polymer

In the isotactic polymer the monomer units have added head-to-tail, to give a series of C–R tertiary bonds with the same configuration in space; in the syndiotactic polymer, the tertiary carbon atoms in the chain have alternate (dextro and levo) spatial configurations; and in the atactic polymer there is no regularity in the distribution of steric configurations of the asymmetric carbon atoms. The various polymeric forms differ in their physical properties. Isotactic polypropylene, for example, has a density of 0.92 g/cm^3 and a melting point of 165° C, whereas an atactic polymer has a somewhat smaller density, 0.85 g/cm^3, and a much lower melting point of −35° C. The more regular isotactic polymer is denser and has a higher melting point than the atactic product because of its greater tendency to crystallize (in spite of the fact that the substituent R may be quite large, hindering crystal formation). Stereoregular polymerization suggests a stereoregulated adsorption at the active centres of the catalyst. In the case of polypropylene, the catalytic centres have been identified by electron micrographs as α-$TiCl_3$ surfaces, which cover only a small fraction of the total surface area, whereas the β-$TiCl_3$ surfaces, which are more abundant, appear to be covered with polymer. The difference between the α- and β-surfaces lies in the random (α) and linear (β) arrangements of Ti^{3+} sites in the two surfaces.

Since the Ziegler–Natta studies, other stereoregulating catalysts have been investigated, notably oxides of chromium, vanadium, molybdenum, and tungsten on silica–alumina or other supports. Other cationic, anionic, and free-radical catalysts are known to produce stereoregulated polymerization. Stereoregular polymerization of dienes has undergone industrial development with the polymerization of isoprene to synthetic natural rubber.

Catalytically active centres

The nature of the active centres in catalytic material is further demonstrated by the enhancement of the catalytic activity of relatively inactive materials when they are subjected to intense radiation. Silica gel bombarded by gamma rays from cobalt-60 turns purplish in colour and becomes capable of inducing the reaction $H_2 + D_2 \rightarrow 2HD$ at liquid nitrogen temperatures. The colour centres, which are positive "holes" (deficiencies) trapped in the vicinity of an oxygen ion next to an aluminum impurity, are bleached *in vacuo* above 200° C and are destroyed by hydrogen even at room temperature.

The properties of dilute concentrations of platinum metals in oxide matrices, such as silica and alumina, as well as on carbon carriers have been studied by Soviet and American scientists. Such catalysts have technical significance in processes for the reforming of gasoline. In such catalysts—containing about 0.5 percent by weight of platinum or palladium—the degree of dispersion of the metal (that is, the ratio of the number of surface metal atoms to the total number present) is close to one. By contrast, on platinum foil the dispersion is only about 4×10^{-3}. The titration and adsorption procedures with hydrogen and oxygen (see above) are employed to evaluate these dispersions.

From these studies it becomes clear that there are two types of behaviour resulting from dispersion. For numerous catalytic processes, ranging from hydrogen–deuterium exchange to the hydrogenation of benzene and the hydrogenolysis of cyclopentane, the reactions are independent of dispersion in the critical region—with catalyst particle size of 50Å or less. Such structure-insensitive processes have been termed facile reactions. On the other hand, there are reactions, such as the isomerization of neopentane to isopentane and simultaneous cracking of the latter to isobutane and methane on platinum–alumina catalysts, where the selectivity for isomerization varies by a factor of 100 for the various catalysts studied (when the hydrogen–neopentane ratio is 10). Thus, the same 1 percent platinum-on-carbon catalyst showed a selectivity ratio of isomerization to hydrogenolysis of 2.5 when the catalyst was reduced in hydrogen at 500° C and a selectivity ratio of 13 when the catalyst was fired *in vacuo* at 900° C (the percentage dispersion remaining at 35 percent in both cases). Such structure-sensitive catalytic reactions have been called demanding reactions. The gain in selectivity appears to be due largely to a reduction in the rate of hydrogenolysis. Since other studies have shown that heating *in vacuo* to 900° tends to develop certain (111) facets of the metal, it is thought that the increase in selectivity is due to a more abundant triadsorption of neopentane on the samples fired at high temperature. It has been shown that a crystallite of platinum about 20Å in size has unusual surfaces not present in a regular octahedral crystallite of similar size. A number of sites where an adsorbed molecule could be surrounded by five platinum nearest neighbours were found on the crystallite with the unusual surface.

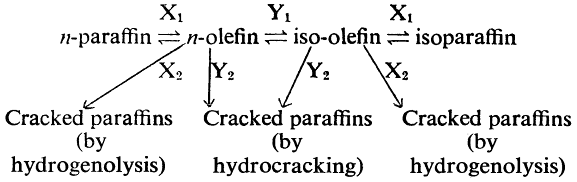
Electronic factors in heterogeneous catalysis

An alternative approach to the problem of surface catalysis involves the consideration of electronic factors in catalyst and reactants. Many catalytic materials are semiconductors. It is thought that these can form a variety of bonds with reactants depending on the free lattice electrons and the "holes" (deficiencies) in the catalyst lattice. Chemisorbed particles react in ways that are dependent on the form of attachment to the surface and that vary with the extent of coverage of the surface as well as with the available supply of electrons and holes. The surface behaves as would free radicals that are introduced directly into the reacting species, dependent on the electrochemical properties of the surface and the bulk of the semiconductor material. Such considerations have led to the determination of the character of the catalyst as a semiconductor and of the adsorbate as an electrochemical species, whether it is composed of positive or negative ions or free atoms or radicals. Catalytic activity also has been explored as a function of the *d*-band character—that is, the number of electrons in *d* orbitals in the atoms of the catalyst materials.

Since 1940 various instrumental techniques have been developed to explore the structure of catalytic materials and the character of the adsorbed species, even during the reaction itself. Among these techniques are electron microscopy, field emission microscopy, electron microprobe methods, magnetic measurements, infrared spectroscopy, Mössbauer spectroscopy, measurements of heats of immersion, flash desorption procedures, low-energy electron diffraction studies, and nuclear magnetic resonance and electron spin resonance techniques.

Polyfunctional heterogeneous catalysis

The term polyfunctional heterogeneous catalysis is applied to a group of catalysts in which more than one component of the surface is active in the processes under study. One example of a bifunctional heterogeneous catalyst is the catalyst of metal (platinum or nickel) deposited on a silica–alumina "acidic" base. Such dual functional catalysts are involved in the interconversions of saturated hydrocarbons (paraffins) and unsaturated hydrocarbons (olefins), and normal (straight-chain) and iso (branched-chain) hydrocarbons, as well as in the splitting (cracking) of the hydrocarbon molecules. The interrelations involved are as follows:



in which X_1 and X_2 are metal-catalyzed processes and Y_1 and Y_2 are acid-catalyzed processes. Operating conditions can be altered to maximize the hydrocracking reactions relative to hydrogenolysis.

A variety of catalysts with "acidic" sites has been found to be active in the dehydration of alcohols and (as above) in the cracking and isomerization of hydrocarbons. Among these are: silica, obtained by calcination (heating) of silica gel; high purity alumina, prepared by the calcining of specially prepared aluminum hydroxide; and silica–alumina mixtures. The catalytic sites have been found to have varying degrees of acidity; their exact nature is still under discussion as well as their characterization in terms of the atomic architecture of the solid catalyst. In the case of silica–alumina, the sites are ascribed to the presence of trivalent aluminum ions Al^{3+}, in a matrix of quadrivalent silicon ions, Si^{4+}, which gives rise to charge differences in the neighbourhood of the aluminum ions. These acidic sites can be poisoned by ammonia and by amines, a finding that confirms their acidic nature. When these catalysts are treated with alkalies their catalytic character is greatly modified. On the other hand, treatment with halogens, especially fluorine or chlorine, enhances the acidic properties of these oxide materials.

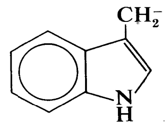
Zeolites

Zeolites are naturally occurring crystalline aluminosilicates having a porous structure and containing cations, generally of the alkali or alkaline earth metals. The cations can be exchanged reversibly with other metal ions without destroying the aluminosilicate structure. Because the zeolites rapidly adsorb certain molecules and exclude others, they have been given the name of molecular sieves. The adsorption characteristics of natural and synthetic zeolites have been studied since the 1930s. Manufactured zeolites, some of which have structures not found in nature, are employed as dehydrating agents but also may be used for the production of catalytic materials by exchange with cationic elements or by impregnation of metal salt solutions into the pores of the zeolite; a large number of zeolitic catalysts have been developed.

Electron donor–acceptor complexes

A class of compounds termed electron donor–acceptor complexes also has been studied for its catalytic activity. The class may be exemplified by a complex between metallic sodium (the donor) and anthracene, $C_{14}H_{10}$, a tricyclic hydrocarbon (the acceptor). The complex can be visualized as an anthracene anion and a sodium cation. Such complexes can exchange the hydrogen of the anion with molecular hydrogen that has been brought into contact with the complex. A complex represented by ZH (in which Z represents all of the molecule except for the exchangeable hydrogen) could undergo exchange with deuterium, as follows: $ZH + D_2 \rightarrow ZD + HD$. It could also take part in corresponding exchanges with hydrocarbons or bring about hydrogenation of hydrocarbons. Among other electron-acceptor catalysts are the metal phthalocyanines (compounds related to certain biological catalysts) and activated charcoal. Some donor–acceptor complexes synthesize ammonia from nitrogen–hydrogen mixtures. This reaction represents a close approach to the activity of biological and bacterial catalysts.

Biological catalysts: the enzymes. Enzymes are substances found in biological systems that are catalysts for specific biochemical processes. Although earlier discoveries of enzymes had been made, a significant confirmation of their importance in living systems was found in 1897 by the German chemist Eduard Buchner, who showed that the filtered cell-free liquor from crushed yeast cells could bring about the conversion of sugar to carbon dioxide. Since that time more than 1,000 enzymes have been recognized, each specific to a particular chemical reaction occurring in living systems. More than 100 of these have been isolated in relatively pure form, including a number of crystallized enzymes. The first enzymes to be crystallized were urease, isolated from the jack bean and crystallized in 1926 by James Batcheller Sumner, and pepsin, crystallized in 1930 by John Howard Northrop, both men being American biochemists. These purified materials were shown to be proteins, chain compounds of about 20 natural amino acids $RCH(NH_2)COOH$, ranging from the simplest, glycine, in which R is hydrogen, to tryptophan, in which R is



Not only have methods been worked out for determining the amino acids, found in an enzyme, but also the sequence of amino acids in an enzyme can be elucidated by a method developed by the English biochemist Frederick Sanger in determining the structure of the protein hormone, insulin. The first enzyme to have its complete amino acid sequence determined in this way was the enzyme bovine pancreatic ribonuclease, which has 124 amino acids in its chain and has a molecular weight of about 14000; the enzyme catalyzes the degradation of ribonucleic acid, a substance active in protein synthesis in living cells. In January 1969, the synthesis of this same enzyme was reported from two different laboratories. The activity of an enzyme depends upon a three-dimensional, or tertiary, structure, but this, in turn, appears to depend solely upon the linear sequence of amino acids. The success of a synthesis of an enzyme can be unequivocally checked by test of its enzymatic activity.

Enzyme reactivity

Enzymes are extremely reactive, as can be shown with a very simple reaction, the splitting of hydrogen peroxide to form water and oxygen, brought about by colloidal metals and by the enzyme catalase. It has been found that one molecule of the latter will cause several million molecules of peroxide to decompose per minute, a rate comparable to that obtained with the best colloidal preparations. This speed of catalase decomposition is probably a maximum for enzymes. Slower acting enzymes normally react at speeds of hundreds of reactions per minute. The rate of reaction is often expressed by an equation developed by L. Michaelis and M.L. Menten of the form:

$$\text{Rate} = \frac{V[S]}{K + [S]}$$

in which V and K are constants for the particular enzymatic process, K being termed the Michaelis constant, and $[S]$ is the concentration of the reactant undergoing change. At low concentrations of S the rate is $V[S]/K$ or proportional to the substrate concentration [S], whereas at high substrate concentrations the $[S]$ terms cancel out and the reaction is essentially independent of the substrate concentration.

A second characteristic of enzymes is their extreme specificity. It has been suggested that each biochemical process has its own specific enzyme. The biochemical processes induced by enzymes fall into broad classifications, such as hydrolysis, decomposition (or "splitting"), synthesis, hydrogenation–dehydrogenation; as with catalysts in general, enzymes are active for both forward and reverse reactions.

Like the laboratory catalysts, enzymes frequently have activators—coenzymes, which may be prosthetic groups (firmly bound to the enzyme itself) and inorganic ions. Adenosine triphosphate, ATP, is an important coenzyme participating in energy-producing processes and passage across cell membranes. Coenzymes often contain vitamins as parts of their structure. Calcium and magnesium ions are important enzyme activators. As with catalysts there are many substances that inhibit, or poison, enzymes. Cyanide ion is a potent inhibitor in many enzymic processes, as are such substances as nerve gases and insecticides. (H.S.Ta.)

Acid–base reactions and equilibria

Acids are chemical compounds that show, in water solution, a sharp taste, a corrosive action on metals, and the ability to turn certain blue vegetable dyes red. Bases are chemical compounds that, in solution, are soapy to the touch and turn red vegetable dyes blue. When mixed, acids and bases neutralize one another and produce salts, substances with a salty taste and none of the characteristic properties of either acids or bases.

Early ideas about acids and bases

The idea that some substances are acids, whereas others are bases, is almost as old as chemistry; and the terms acid, base, and salt occur very early in the writings of the medieval alchemists. Acids were probably the first of these to be recognized, apparently because of their sour taste. The English word acid, the French *acide,* the German *Säure,* and the Russian *kislota* are all derived from words meaning sour (Latin *acidus,* German *sauer,* Old Norse *sūur,* and Russian *kisly*). Other properties associated at an early date with acids were their solvent, or corrosive, action; their effect on vegetable dyes; and the effervescence resulting when they were applied to chalk (production of bubbles of carbon dioxide gas). Bases (or alkalies) were characterized mainly by their ability to neutralize acids and form salts, the latter being typified rather loosely as crystalline substances soluble in water and having a saline taste.

In spite of their imprecise nature, these ideas served to correlate a considerable range of qualitative observations; and many of the commonest chemical materials that early chemists encountered could be classified as acids (hydrochloric, sulfuric, nitric, and carbonic acids), bases (soda, potash, lime, ammonia) or salts (common salt, sal ammoniac, saltpetre, alum, borax). The absence of any apparent physical basis for the phenomena concerned made it difficult to make quantitative progress in understanding acid–base behaviour, but the ability of a fixed quantity of acid to neutralize a fixed quantity of base was one of the earliest examples of chemical equivalence: the idea that a certain measure of one substance is in some chemical sense equal to a different amount of a second substance. In addition, it was found quite early that one acid could be displaced from a salt with another acid, and this made it possible to arrange acids in an approximate order of strength. It also soon became clear that many of these displacements could take place in either direction according to experimental conditions. This phenomenon suggested that acid-base reactions are reversible—that is, that the products of the reaction can interact to regenerate the starting material. It also introduced the concept of equilibrium to acid–base chemistry—this concept states that reversible chemical reactions reach a point of balance, or equilibrium, at which the starting materials and the products are each regenerated by one of the two reactions as rapidly as they are consumed by the other.

Apart from their theoretical interest, acids and bases play a large part in industrial chemistry and in everyday life. Sulfuric acid and sodium hydroxide are among the products of chemical industry manufactured in largest amount, and a large percentage of chemical processes involve acids or bases as reactants or as catalysts. Almost every biological chemical process is closely bound up with acid–base equilibria in the cell, or in the organism as a whole, and the acidity or alkalinity of the soil and water are of great importance for the plants or animals living in them. Both the ideas and the terminology of acid–base chemistry have permeated daily life, and the term salt is especially common.

THEORETICAL DEFINITIONS OF ACIDS AND BASES

The oxygen theory of acids

Hydrogen and hydroxide ions. The first attempt at a theoretical interpretation of acid behaviour was made by Antoine-Laurent Lavoisier at the end of the 18th century. Lavoisier supposed that all acids must contain oxygen, and this idea was incorporated in the names used for this element in the various languages; the English oxygen, from the Greek *oxys* (sour) and *genna* (production); the German *Sauerstoff,* literally acid material; and the Russian *kislorod,* from *kislota* (acid). Following the discovery that hydrochloric acid contained no oxygen, Sir Humphry Davy in about 1815 first recognized that the key element in acids was hydrogen. Not all substances that contain hydrogen, however, are acids, and the first really satisfactory definition of an acid was given by Justus von Liebig of Germany in 1838. According to Liebig, an acid is a compound containing hydrogen in a form in which it can be replaced by a metal. This definition held the field for about 50 years and is still considered essentially correct, though somewhat outmoded. At the time of Liebig's proposal, bases were still regarded solely as substances that neutralized acids with the production of salts, and nothing was known about the constitutional features of bases that enabled them to do this.

The whole subject of acid–base chemistry acquired a new look and a quantitative aspect with the advent of the electrolytic dissociation theory propounded by Wilhelm Ostwald and Svante August Arrhenius (both Nobel laureates) in the 1880s. The principal feature of this theory is that certain compounds, called electrolytes, dissociate in solution to give ions. With the development of this theory it was realized that acids are merely those hydrogen compounds that give rise to hydrogen ions (H^+) in aqueous solution. It was also realized at that time that there is a correspondence between the degree of acidity of a solution (as shown by effects on vegetable dyes and other properties) and the concentration of hydrogen ions in the solution. Correspondingly, basic (or alkaline) properties could then be associated with the presence of hydroxide ions (OH^-) in aqueous solution, and the neutralization of acids by bases could be explained in terms of the reaction of these two ions to give the neutral molecule water ($H^+ + OH^- \rightarrow H_2O$.) This led naturally to the following simple definitions: Acids and bases are substances that give rise, respectively, to hydrogen and hydroxide ions in aqueous solution. This definition was generally accepted for the next 30 or 40 years. In purely qualitative terms, it does not offer many advantages over Liebig's definition of acids, but it does provide a satisfactory definition for bases.

Hydrogen and hydroxide ion definitions of acids and bases

Nevertheless, there is a great advantage in the definition of acids and bases in terms of hydrogen and hydroxide ions and this advantage lies in its quantitative aspects. Because the concentrations of hydrogen and hydroxide ions in solution can be measured, notably by determining the electrical conductivity of the solution (its ability to carry an electrical current), a quantitative measure of the acidity or alkalinity of the solution is provided. Moreover, the equations developed to express the relationships between the various components of reversible reactions can be applied to acid and base dissociations to give definite values, called dissociation constants. These constants can be used to characterize the relative strengths (degrees of dissociation) of acids and bases and, for this reason, supersede earlier semiquantitative estimates of acid or base strength. As a result of this approach, a satisfactory quantitative description was given at an early date of a large mass of experimental observations, a description that remains essentially unaffected by later developments in definitions of acid–base reactions.

The success of these quantitative developments, however, unfortunately helped to conceal some ambiguities and logical inconsistencies in the qualitative definitions of acids and bases in terms of the production of hydrogen and hydroxide ions, respectively. For example, it was not clear whether a substance like anhydrous hydrogen chloride, which would not conduct electricity, should be regarded as an acid, or whether it should be considered an acid only after it had come in contact with water. A modified definition of bases also seemed to be required that would be applicable to nonaqueous solutions, in which the anion (negatively charged ion) produced is not the hydroxide ion, as is the case in water, but varies from solvent to solvent, the methoxide ion (CH_3O^-) acting as the basic anion in methanol (CH_3OH), for example, and the amide ion (NH_2^-) playing the same role in liquid ammonia (NH_3). Even with acids the solvent is involved, since there is much evidence to show that the so-called hydrogen ion in solution does not exist as H^+ but always contains at least one molecule of solvent, as H_3O^+ in water, $CH_3OH_2^+$ in methanol, and NH_4^+ in liquid ammonia. These considerations led to the development of definitions of acids and bases that depended on the solvent (see below). In spite of this change, however, the difficulty still remained that typical acid–base properties, such as neutralization, indicator (vegetable-dye) effects, and catalysis, often took place in solvents such as benzene or chloroform in which free ions could barely be detected at all (by conductivity measurements). Even for aqueous solutions a particular ambiguity arises in the definition of bases, some of which (for example, metallic hydroxides) contain a hydroxyl group, whereas others (such as amines) do not. The latter produce hydroxide ions in solution by reacting with water molecules.

The Brønsted–Lowry definition. In order to resolve the various difficulties in the hydrogen–hydroxide ion definitions of acids and bases, a new, more generalized definition was proposed in 1923 almost simultaneously by J.M. Brønsted and T.M. Lowry. Although the pursuit of exact verbal definitions of qualitative concepts is usually not profitable in physical science, the Brønsted–Lowry definition of acids and bases has had far-reaching consequences in the understanding of a wide range of phenomena and in the stimulation of much experimental work. The definition is as follows: an acid is a species having a tendency to lose a proton, and a base is a species having a tendency to gain a proton. The term proton means the species H^+ (the nucleus of the hydrogen atom) rather than the actual hydrogen ions that occur in various solutions; the definition is thus independent of the solvent. The use of

Proton definition of acids and bases

the word species rather than substance or molecule implies that the terms acid and base are not restricted to uncharged molecules but apply also to positively or negatively charged ions. This extension is one of the important features of the Brønsted–Lowry definition. It can be summarized by the equation $A \rightleftarrows B + H^+$ in which A and B together are a conjugate acid–base pair. In such a pair A must obviously have one more positive charge (or one less negative charge) than B, but there is no other restriction on the sign or magnitude of the charges. Several examples of conjugate acid–base pairs are given in Table 1.

Table 1: Examples of Conjugate Acid-Base Pairs

acid	base
Acetic acid, CH_3CO_2H	acetate ion, $CH_3CO_2^-$
Bisulfate ion, HSO_4^-	sulfate ion, SO_4^{2-}
Ammonium ion, NH_4^+	ammonia, NH_3
Ammonia, NH_3	amide ion, NH_2^-
Water, H_2O	hydroxide ion, OH^-
Hydronium (oxonium) ion, H_3O^+	water, H_2O

A number of points about the Brønsted–Lowry definition should be emphasized:

1. As mentioned above, this definition is independent of the solvent. The ions derived from the solvent (H_3O^+ and OH^- in water, and NH_4^+ and NH_2^- in liquid ammonia) are not accorded any special status but appear as examples of acids or bases in terms of the general definition. On the other hand, of course, they will be particularly important species for reactions in the solvent to which they relate.

2. In addition to the familar molecular acids, two classes of ionic acids emerge from the new definition. The first comprises anions derived from acids containing more than one acidic hydrogen; *e.g.*, the bisulfate ion (HSO_4^-, as in Table 1) and primary and secondary phosphate ions ($H_2PO_4^-$ and HPO_4^{2-}) derived from phosphoric acid (H_3PO_4). The second and more interesting class consists of positively charged ions (cations), such as the ammonium ion (NH_4^+), which can be derived by the addition of a proton to a molecular base, in this case ammonia (NH_3). The hydronium ion (H_3O^+), which is the hydrogen ion in aqueous solution, also belongs to this class. The charge of these ionic acids, of course, always must be balanced by ions of opposite charges, but these oppositely charged ions usually are irrelevant to the acid-base properties of the system. For example, if sodium bisulfate ($Na^+HSO_4^-$) or ammonium chloride ($NH_4^+Cl^-$) is used as an acid, the sodium ion (Na^+) and the chloride ion (Cl^-) contribute nothing to the acidic properties and could equally well be replaced by other ions, such as potassium (K^+) and perchlorate (ClO_4^-), respectively.

3. Molecules such as ammonia and organic amines are bases by virtue of their tendency to accept a proton. With metallic hydroxides such as sodium hydroxide, on the other hand, the basic properties are due to the hydroxide ion itself, the sodium ion serving merely to preserve electrical neutrality. Moreover, not only the hydroxide ion, but also the anions of other weak acids (for example the acetate ion) must be classed as bases because of their tendency to reform the acid by accepting a proton. Formally, the anion of any acid might be regarded as a base, but for the anion of a very strong acid (the chloride ion, for example) the tendency to accept a proton is so weak that its basic properties are insignificant and it is inappropriate to describe it as a base. Similarly, all hydrogen compounds could formally be defined as acids, but in many of them (for example, most hydrocarbons, such as methane, CH_4) the tendency to lose a proton is so small that the term acid would not normally be applied to them.

Amphoteric substance

4. Some species, including molecules as well as ions, possess both acidic and basic properties; such materials are said to be amphoteric. As shown in Table 1, both water and ammonia are amphoteric, a situation that can be represented by the schemes H_3O^+–H_2O–OH^- and NH_4^+–NH_3–NH_2^-. Another example is the secondary phosphate ion, HPO_4^{2-}, which can either lose or accept a proton, according to the following equations: $HPO_4^{2-} \rightleftarrows PO_4^{3-} + H^+$ and $HPO_4^{2-} + H^+ \rightleftarrows H_2PO_4^-$. The amphoteric properties of water are particularly important in determining its properties as a solvent for acid–base reactions.

5. The equation $A \rightleftarrows B + H^+$, used in the Brønsted–Lowry definition, does not represent a reaction that can be observed in practice, since the free proton, H^+, can be observed only in gaseous systems at low pressures. In solution, the proton always is attached to some other species, commonly a solvent molecule. Thus in water the ion H_3O^+ consists of a proton bound to a water molecule. For this reason all observable acid–base reactions in solution are combined in pairs, with the result that they are of the form $A_1 + B_2 \rightleftarrows B_1 + A_2$. The fact that the process $A \rightleftarrows B + H^+$ cannot be observed does not imply any serious inadequacy of the definition. A similar situation exists with the definitions of oxidizing and reducing agents, which are defined respectively as species having a tendency to gain or lose electrons, even though one of these reactions never occurs alone and free electrons are never detected in solution (any more than free protons are).

Alternative definitions. Although the Brønsted–Lowry concept of acids and bases as donors and acceptors of protons is still the most generally accepted one, other definitions are often encountered. Certain of these are adapted for special situations only, but the most important of these other definitions is in some respects more general than the Brønsted–Lowry definition. This definition was first proposed by the U.S. chemist Gilbert N. Lewis in 1923.

The Lewis definition of acids and bases

According to Lewis, an acid is a species that can accept an electron pair from a base with the formation of a chemical bond composed of a shared electron pair (covalent bond). This classification includes as bases the same species covered by the Brønsted–Lowry definition, since a molecule or ion that can accept a proton does so because it has one or more unshared pairs of electrons, and therefore it also can combine with electron acceptors other than the proton. On the other hand, the typical Lewis acids need not (and usually do not) contain protons, being species with outer electron shells that are capable of expansion, such as boron trifluoride (BF_3), sulfur trioxide (SO_3), and silver ion (Ag^+). Lewis originally based his ideas on the experimental fact that these nonprotonic acids often exhibit the properties regarded as typical of acids, such as neutralization of bases, action on indicators, and catalysis. Such substances often are electron acceptors, but this is not always the case; carbon dioxide (CO_2) and nitrogen pentoxide (N_2O_5), for example, contain completed octets of electrons and, according to usual valence theory, cannot accept any more. In addition, hydrogen-containing substances that have always been regarded as acids (acetic acid, for example) are not obviously electron acceptors, being rather adducts of the proton (a true Lewis acid) and a base such as the acetate ion. They can only be brought logically into the Lewis scheme by appealing to the fact that the reaction between a proton acid, which may be designated as XH, and a base, denoted by B, passes through an intermediate hydrogen-bonded state, X-H . . . B (in which the dotted line indicates a hydrogen bond, a relatively weak secondary attractive force).

Numerous lengthy polemical exchanges have taken place regarding the relative merits of the Brønsted–Lowry and Lewis definitions of acids and bases. The difference is essentially one of nomenclature and has little scientific content. In the remainder of this article the term acid will be used to denote a proton donor (following the Brønsted–Lowry terminology) whereas the term Lewis acid will be employed exclusively to refer to electron-pair acceptors. This choice is based partly upon the logical difficulties mentioned in the last paragraph, and partly on the fact (see below) that the quantitative description of acid–base reactions is much simpler when it is confined to proton acids. This terminology also represents the commonest usage of the terms.

The definition of Lewis acids and bases in terms of the gain or loss of electrons should not be confused with the definition of oxidizing and reducing agents in similar terms. In oxidation–reduction reactions one or more electrons is transferred completely from the reducing agent to the oxidizing agent, whereas in a Lewis acid–base reaction an electron pair on the base is used to form a covalent link with the acid.

Generalized solvent definitions

Certain other acid–base definitions have been based upon reactions occurring in specific solvent systems. For proton acids in amphoteric solvents these are equivalent to the Brønsted–Lowry definition. It is sometimes convenient to have general terms for the cation and anion derived from the solvent molecule by the additon and removal of a proton, respectively. The terms lyonium and lyate ions are occasionally used in this way. In water, the lyonium and lyate ions are H_3O^+ and OH^-; in ethanol, they are $C_2H_5OH_2^+$ and $C_2H_5O^-$; and in liquid ammonia, NH_4^+ and NH_2^-. For a given solvent, an acid can then be defined as a substance that increases the lyonium ion concentration (and correspondingly decreases the lyate ion concentration), whereas a base increases the lyate ion concentration (and decreases the lyonium ion concentration). This kind of definition, to be sure, really does not add anything to the concept of acids and bases as proton donors and proton acceptors.

The idea that an acid is a solute that gives rise to cations characteristic of the solvent, and a base is a solute that gives rise to anions characteristic of the solvent, has sometimes been extended to solvents where no protons are involved at all, for example, liquid sulfur dioxide, SO_2. In this example, the solvent is supposed to ionize according to the equation $2SO_2 \rightleftarrows SO^{2+} + SO_3^{2-}$. Thionyl chloride, regarded as $SO^{2+} + 2Cl^-$, then can be considered an acid, and potassium sulfite, $2K^+ + SO_3^{2-}$, a base. The species SO^{2+} and SO_3^{2-} can certainly be regarded as Lewis acids and bases, but it is doubtful that they exist to any appreciable extent in liquid sulfur dioxide, a situation that makes the discussion somewhat artificial. Although this view of acids and bases has been useful in stimulating work in unusual types of solvent (for example in carbonyl chloride, selenium oxychloride, antimony trichloride, and hydrogen cyanide) it has not met with general acceptance.

ACID–BASE REACTIONS

Proton-transfers. As already mentioned, the reaction expressed by the Brønsted–Lowry definition, $A \rightleftarrows B + H^+$, does not actually occur in any solution processes. This is because H^+, the bare proton, has an enormous tendency to add to almost all chemical species and cannot exist in any detectable concentrations except in a high vacuum. Apart from any specific chemical interaction, the very small size of the proton (about 10^{-15} metres) means that it exerts an extremely powerful electric field, which will polarize and therefore attract any molecule or ion it comes into contact with. It has been estimated that the dissociation of 19 grams of the hydronium ion H_3O^+ to give 1 gram of protons and 18 grams of water would require the expenditure of about 1,200,000 joules (290,000 calories) of energy, and thus it is an extremely unlikely process indeed.

Combined pairs of acid–base reactions

Typical acid–base reactions may be thought of as the combination of two reaction schemes $A_1 \rightleftarrows B_1 + H^+$ and $H^+ + B_2 \rightleftarrows A_2$, leading to the combined form: $A_1 + B_2 \rightleftarrows B_1 + A_2$. This represents a proton-transfer reaction from A_1 to B_2, producing B_1 and A_2. A large number of reactions in solution, often referred to under a variety of names, can be represented in this way. This is illustrated by the following examples, in each of which the species are written in the order A_1, B_2, B_1, A_2.

Dissociation of molecular acids in water. In this instance, water acts as a base. The equation for the dissociation of acetic acid, for example, is as follows: $CH_3CO_2H + H_2O \rightleftarrows CH_3CO_2^- + H_3O^+$.

Dissociation of bases in water. In this case, the water molecule acts as an acid and adds a proton to the base. An example, using ammonia as the base, is $H_2O + NH_3 \rightleftarrows OH^- + NH_4^+$. Older formulations would have written the left-hand side of the equation as ammonium hydroxide, NH_4OH, but it is not now believed that this species exists, except as a weak, hydrogen-bonded complex.

Dissociation of acids and bases in nonaqueous solvents. These situations are entirely analogous to the comparable reactions in water. For example, the dissociation of acetic acid in methanol may be written as $CH_3CO_2H + CH_3OH \rightleftarrows CH_3CO_2^- + CH_3OH$, and the dissociation of ammonia in the same solvent as $CH_3OH + NH_3 \rightleftarrows CH_3O^- + NH_4^+$.

Self-dissociation of amphoteric solvents. In this case, one solvent molecule acts as an acid, and another as a base. Self-dissociation of water and liquid ammonia may be given as examples:

$$H_2O + H_2O \rightleftarrows OH^- + H_3O^+$$

$$NH_3 + NH_3 \rightleftarrows NH_2^- + NH_4^+.$$

Neutralization. For a strong acid and a strong base in water, the neutralization reaction is between hydrogen and hydroxide ions; *i.e.,* $H_3O^+ + OH^- \rightleftarrows 2H_2O$. For a weak acid and a weak base neutralization is more appropriately considered to involve direct proton transfer from the acid to the base. For example, the neutralization of acetic acid by ammonia may be written as $CH_3CO_2H + NH_3 \rightarrow CH_3CO_2^- + NH_4^+$. This equation does not involve the solvent; it therefore also represents the process of neutralization in an inert solvent, such as benzene, or in the complete absence of a solvent. (If one of the reactants is present in large excess, the reaction is more appropriately described as the dissociation of acetic acid in liquid ammonia, or of ammonia in glacial acetic acid.)

Acidic and basic salt solutions

Hydrolysis of salts. Many salts give aqueous solutions with acidic or basic properties. This is termed hydrolysis, and the explanation of hydrolysis reactions in classical acid–base terms was somewhat involved. In terms of the Brønsted–Lowry concept, however, hydrolysis appears to be a natural consequence of the acidic properties of cations derived from weak bases and the basic properties of anions derived from weak acids. For example, hydrolysis of aqueous solutions of ammonium chloride and of sodium acetate is represented by the following equations:

$$(Cl^-) + NH_4^+ + H_2O \rightleftarrows (Cl^-) + NH_3 + H_3O^+$$

$$(Na^+) + H_2O + CH_3CO_2^- \rightleftarrows (Na^+) + OH^- + CH_3CO_2H$$

The sodium and chloride ions take no part in the reaction and could equally well be omitted from the equations.

The acidity of the solution represented by the first equation is due to the presence of the hydronium ion (H_3O^+) and the basicity of the second comes from the hydroxide ion (OH^-). The reverse reactions simply represent, respectively, the neutralization of aqueous ammonia by a strong acid and of aqueous acetic acid by a strong base.

A superficially different type of hydrolysis occurs in aqueous solutions of salts of some metals, especially those giving multiply charged cations. For example, aluminum, ferric, and chromic salts all give aqueous solutions that are acidic. This behaviour also can be interpreted in terms of proton-transfer reactions if it is remembered that the ions involved are strongly hydrated in solution. In a solution of an aluminum salt, for instance, a proton is transferred from one of the water molecules in the hydration shell to a molecule of solvent water. The resulting hydronium ion (H_3O^+) accounts for the acidity of the solution:

$$Al(H_2O)_6^{3+} + H_2O \rightleftarrows Al(H_2O)_5OH^{2+} + H_3O^+.$$

Formation of adducts

Reactions of Lewis acids. In the reaction of a Lewis acid with a base the essential process is the formation of an adduct in which the two species are joined by a covalent bond; proton transfers are not normally involved. If both the Lewis acid and base are uncharged, the resulting bond is termed semipolar or coordinate, as in the reaction of boron trifluoride with ammonia.

$$\underset{\text{acid}}{BF_3} + \underset{\text{base}}{NH_3} \rightarrow \underset{\text{adduct.}}{F_3\overset{-}{B}{-}\overset{+}{N}H_3}$$

Frequently, however, either or both species bears a charge (most commonly a positive charge on the acid or a negative charge on the base), and the location of charges within the adduct often depends upon the theoretical interpretation of the valences involved. Examples of such reactions are:

$$\underset{\text{silver ion}}{Ag^+} + \underset{\text{ammonia}}{2NH_3} \rightarrow \underset{\text{adduct}}{[Ag(NH_3)_2]^+}$$

$$\underset{\text{boron trifluoride}}{BF_3} + \underset{\text{hydroxide ion}}{OH^-} \rightarrow \underset{\text{adduct}}{[BF_3(OH)]^-}$$

$$\underset{\text{cadmium ion}}{Cd^{2+}} + \underset{\text{cyanide ion}}{4CN^-} \rightarrow \underset{\text{adduct.}}{[Cd(CN)_4]^{2-}}$$

In another common type of process, one acid, or base, in an adduct is replaced by another:

$$\underset{\text{adduct (1)}}{Cl_4\overset{-}{Sn}{-}\overset{+}{N}H_3} + \underset{\text{acid (2)}}{BF_3} \rightarrow \underset{\text{adduct (2)}}{F_3\overset{-}{B}{-}\overset{+}{N}H_3} + \underset{\text{acid (1)}}{SnCl_4}$$

$$\underset{\text{adduct (1)}}{Al(H_2O)_6^{3+}} + \underset{\text{base (2)}}{6F^-} \rightarrow \underset{\text{adduct (2)}}{AlF_6^{3-}} + \underset{\text{base (1).}}{6H_2O}$$

In fact, reactions such as the simple adduct formations above often are formulated more correctly as replacements. For example, if the reaction of boron trifluoride with ammonia is carried out in ether as a solvent, it becomes a replacement reaction:

$$(C_2H_5)_2\overset{+}{O}{-}\overset{-}{B}F_3 + NH_3 \rightarrow F_3\overset{-}{B}{-}\overset{+}{N}H_3 + (C_2H_5)_2O.$$

Similarly, the reaction of silver ions with ammonia in aqueous solution is better written as a replacement reaction:

$$Ag(H_2O)_2^+ + 2NH_3 \rightarrow Ag(NH_3)_2^+ + 2H_2O.$$

Furthermore, if most covalent molecules are regarded as adducts of (often hypothetical) Lewis acids and bases, an enormous number of reactions can be formulated in the same way. To take a single example, the reaction of methyl chloride with hydroxide ion to give methanol and chloride ion (usually written as $CH_3Cl + OH^- \rightarrow CH_3OH + Cl^-$) can be reformulated as replacement of a base in a Lewis acid–base adduct, as follows: (adduct of CH_3^+ and Cl^-) + OH^- → (adduct of CH_3^+ and OH^-) + Cl^-. Opinions differ as to the usefulness of this extremely generalized extension of the Lewis acid–base-adduct concept.

Reactions of anhydrous oxides

The reactions of anhydrous oxides (usually solid or molten) to give salts may be regarded as examples of Lewis acid–base-adduct formation. For example, in the reaction of calcium oxide with silica to give calcium silicate, the calcium ions play no essential part in the process, which may be considered therefore to be adduct formation between silica as the acid and oxide ion as the base:

$$\underset{\text{acid}}{SiO_2} + \underset{\text{base}}{O^{2-}} \rightarrow \underset{\text{adduct.}}{SiO_3^{2-}}$$

A great deal of the chemistry of molten-oxide systems can be represented in this way, or in terms of the replacement of one acid by another in an adduct.

Acid–base catalysis. Acids (including Lewis acids) and bases act as powerful catalysts for a great variety of chemical reactions, in the laboratory, in industry, and in processes occurring in nature. Historically, catalytic action was regarded as one of the essential characteristics of acids, and the parallel occurrence of catalytic action and electrical conductivity was one of the compelling pieces of evidence in establishing the theory of electrolytic dissociation as the basis of acid–base behaviour at the end of the 19th century.

Acid–base catalysis was originally thought of in terms of a mysterious influence of the acid or base, but it is now generally believed to involve an actual acid–base reaction between the catalyst and the reacting substance, termed the substrate, with the catalyst being regenerated at a later stage of the reaction. Moreover, knowledge of reaction mechanisms is now sufficient to suggest detailed sequences of reactions for many acid- or base-catalysis reactions, most of which sequences are at least plausible and in many instances are well established.

Regeneration of the catalyst

In most acid–base reactions the addition or removal of a proton does not bring about any drastic change in the structure of the molecule, or in its stability or reactivity. It is a characteristic of reactions catalyzed by acids or bases, however, that the addition or removal of a proton either makes the substrate unstable, so that it decomposes or rearranges, or that it causes the substrate to become reactive toward some other species present in the system. In cases of rearrangement, the regeneration of the catalyst often involves the removal or addition of a proton at a site other than that at which the initial addition or removal took place. It is not necessary that the substrate in an acid- or base-catalyzed reaction should itself have marked acid–base properties, since even a very small extent of initial acid–base reaction may be enough to bring about the subsequent change.

Instances of acid–base catalysis are numerous indeed; a few examples are given here, as follows:

Isomerization of olefins, acid-catalyzed. Unsaturated compounds frequently rearrange reversibly under the influence of acids to give products in which the double bond occurs in a new location. The interconversion of 2-butene and 1-butene is shown here:

$$CH_3CH = CHCH_3 + H_3O^+ \rightleftarrows CH_3CH_2\overset{+}{C}HCH_3 + H_2O$$

$$CH_3CH_2\overset{+}{C}HCH_3 + H_2O \rightleftarrows CH_3CH_2CH = CH_2 + H_3O^+.$$

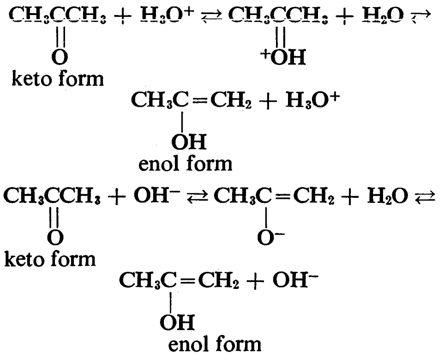
Reversible dehydration of alcohols, acid-catalyzed. Under the influence of acids, alcohols generally undergo loss of water to give olefinic products. The dehydration of ethanol to ethylene occurs as follows:

$$CH_3CH_2OH + H_3O^+ \rightleftarrows CH_3CH_2OH_2^+ + H_2O$$

$$CH_3CH_2OH_2^+ \rightleftarrows CH_3CH_2^+ + H_2O \rightleftarrows$$

$$CH_2 = CH_2 + H_3O^+.$$

Keto–enol tautomerism, acid- and base-catalyzed. Acids and bases both bring about the establishment of an equilibrium between ketones (or aldehydes) and their enol forms, which contain a hydroxyl group directly attached to a doubly bonded carbon atom ($-\underset{\mathrm{OH}}{\underset{|}{C}} = C-$). The interconversion between the two forms is called keto–enol tautomerism. The reaction cannot always be observed directly, since the enol form may not reach measurable concentrations, even at equilibrium, but the highly active enol may be detected by its reaction with various reagents, notably the halogens (bromine, for example). Keto–enol tautomerization of acetone can be brought about by acid or base catalysis, as follows:



$$\underset{\text{keto form}}{CH_3\underset{O}{\underset{\|}{C}}CH_3} + H_3O^+ \rightleftarrows CH_3\underset{^+OH}{\underset{\|}{C}}CH_3 + H_2O \rightleftarrows \underset{\text{enol form}}{CH_3\underset{OH}{\underset{|}{C}}{=}CH_2} + H_3O^+$$

$$\underset{\text{keto form}}{CH_3\underset{O}{\underset{\|}{C}}CH_3} + OH^- \rightleftarrows CH_3\underset{O^-}{\underset{|}{C}}{=}CH_2 + H_2O \rightleftarrows \underset{\text{enol form}}{CH_3\underset{OH}{\underset{|}{C}}{=}CH_2} + OH^-$$

Aldol condensation, base-catalyzed. Self-condensation of aldehydes, the so-called aldol condensation, occurs readily, when catalyzed by bases, to give β-hydroxy aldehydes. The prototype of this reaction is the conversion of acetaldehyde to β-hydroxybutyraldehyde, or aldol. The first step of this reaction is the production of an enolate ion (as in formation of the keto–enol tautomeric mixture), but this anion then reacts with a second molecule of acetaldehyde to give the product as shown below:

$$CH_3CHO + OH^- \rightleftarrows [CH_2CHO]^- + H_2O$$

$$CH_3CHO + [CH_2CHO]^- \rightleftarrows CH_3CH(O^-)CH_2CHO$$

$$CH_3CH(O^-)CH_2CHO + H_2O \rightleftarrows \underset{\text{aldol}}{CH_3CHOHCH_2CHO} + OH^-$$

Importance of acid–base catalysis

These examples illustrate the importance of acid–base catalysis in organic reactions. The equations have been written in terms of H_3O^+ and OH^- as the acid and base catalysts, respectively, and these are certainly the most important catalysts in aqueous solution. For many of these reactions (especially isomerization of olefins and dehydration of alcohols), there is ample evidence that other acids or bases also can act as catalysts. This behaviour is known as general acid–base catalysis. It appears particularly clearly in inert solvents such as benzene, in which catalysis by molecular acids and bases is frequently observed despite the absence of detectable quantities of ions derived from the solvent. Acidic groups, such as sulfonic acid ($-SO_3H$) and carboxylic acid groups ($-CO_2H$), attached to a solid molecular framework (as in some ion exchange resins) also act as heterogeneous catalysts for many chemical reactions.

The above examples show that proton-transfer processes can play a specific part in reaction mechanisms and, in these and similar instances, it is doubtful whether any uncatalyzed or spontaneous reaction of the same type can take place. Apparent evidence to the contrary can usually be explained by catalysis by solvent molecules or by adventitious acidic or basic impurities.

Lewis-acid catalysis

Lewis acids can exert a catalytic effect in two different ways. In the first of these they interact with hydrogen-containing compounds present in the system to assist the release of a proton to the substrate. For example, the polymerization of olefins by Lewis acids, such as boron trifluoride (BF_3), aluminum chloride ($AlCl_3$), and titanium tetrachloride ($TiCl_4$), is believed to be caused by their interaction with proton acids (for example, traces of water) and the olefin to give a carbonium ion, which then reacts further with more olefin:

$$BF_3 + H_2O + CHR = CHR' \rightarrow CH_2R{-}\overset{+}{C}HR' + [BF_3OH]^-$$

$$CH_2R{-}\overset{+}{C}HR' + CHR = CHR' \rightarrow CH_2RCHR'CHR\overset{+}{C}HR', \text{ etc.}$$

In the second mode of action, the Lewis acid acts directly on the substrate, and by withdrawing electrons converts it into a reactive form. A typical example is the action of catalysts like aluminum chloride on alkyl halides to produce carbonium ions: $RCl + AlCl_3 \rightarrow R^+ + [AlCl_4]^-$. The carbonium ion can then react further with other substances, for example, aromatic hydrocarbons. The same type of catalysis probably occurs with many solid oxide catalysts (for example, aluminosilicates), although it is often difficult to decide whether the catalytic action of these materials is due to centres with a deficiency of electrons or to acidic hydroxyl groups.

ACID–BASE EQUILIBRIA

Certain general principles apply to any solvent with both acidic and basic properties, for example, water, alcohols, ammonia, amines, and acetic acid. Denoting the solvent molecule by SH, proton transfer can give rise to the ions SH_2^+ and S^-, sometimes called lyonium and lyate ions, respectively (see above). In the pure solvent these are the only ions present, and they must be present in equal concentrations to preserve electrical neutrality. The equilibrium involved, therefore, is as follows: $2SH \rightleftarrows SH_2^+ + S^-$.

Equilibrium constants

The equilibrium constant (K_s') for this reaction (the mathematical quantity that expresses the relationships between the concentrations of the various species present at equilibrium) would normally be given by the equation $K_s' = [SH_2^+]\ [S^-]/[SH]^2$, in which the square brackets denote the concentrations of the species within the brackets. In a given solvent, however, the concentration of the solvent, [SH], is a large and constant quantity, and it is therefore usual to eliminate this term and express the self-dissociation of the solvent by the equation $K_s = [SH_2^+][S^-]$. In this equation, K_s is termed the ion product or the autoprotolysis constant of the solvent. The concentrations are usually expressed in moles per litre, a mole being the molecular weight of the compound in grams. Since a solvent that is a good proton donor is normally a poor proton acceptor, and vice versa, the degree of ionization is generally low and K_s is usually a small quantity. It is about 10^{-14} for water at ordinary temperatures, and one of the largest K_s values known is 1.7×10^{-4} for 100 percent sulfuric acid. The above equation applies not only to the pure solvent, but also (with the same value of K_s) to any dilute solutions of acids, bases, or salts in the solvent in question. In these solutions $[SH_2^+]$ and $[S^-]$ need not be equal, since the condition of electrical neutrality involves the concentration of other ions as well, and it is obvious from the equation that a high value of $[SH_2^+]$ must imply a low value of $[S^-]$, and vice versa.

If an acid A is added to the solvent SH it will be at least partly converted into the conjugate base B according to the reaction $A + SH \rightleftarrows B + SH_2^+$, which would be characterized formally by an equilibrium constant $[B][SH_2^+]/[A]\ [SH]$. Again, however, it is usual to omit the term for the constant concentration of the solvent, [SH], from this expression, and to define a constant K_a by the equation

$$\frac{[B]\ [SH_2^+]}{[A]} = K_a,$$

Dissociation constants

which is known as the dissociation constant of the acid A in the solvent SH. Any acid–base reaction $A_1 + B_2 \rightleftarrows B_1 + A_2$ will proceed from left to right almost completely if A_1 is a much stronger acid than A_2. It is a natural extension of this idea to use the equilibrium constant as a measure of the strength of A_1 relative to A_2. The dissociation constant is thus (apart from the constant factor [SH], which has been omitted) a measure of the acid strength of A relative to that of the lyonium ion SH_2^+.

In some instances reaction goes so completely from left to right that it is not possible to measure the equilibrium constant. A is said then to be a strong acid in the solvent SH; similarly, acids with readily measurable dissociation constants (in practice less than about 0.1) are known as weak acids.

Similar considerations apply to solutions of bases. The reaction involved in this case is $SH + B \rightleftarrows S^- + A$ and the equilibrium constant K_b defined by

$$\frac{[S^-]\ [A]}{[B]} = K_b$$

is known as the dissociation constant of the base B. Apart from the omitted constant factor [SH], K_b represents the basic strength of B relative to that of the lyate ion S^-. Bases are termed strong and weak in the same way that acids are.

The values of K_a and K_b for a conjugate acid–base pair A–B in a given solvent are not independent, since consideration of the dissociation constants of the solvent, acid, and base show that $K_aK_b = [SH_2^+][S^-] = K_s$ in which K_s is the ion product of the solvent. It is therefore unnecessary to specify both K_a and K_b, and it has become common practice to characterize an acid–base pair by K_a only, which may be termed the acidity constant of A–B in the solvent SH. If the value of K_b is required it is readily obtained from K_a and K_s. Since readily accessible values of K_a are always much less than unity, it is often convenient to introduce a quantity pK_a, sometimes called the acidity exponent, and defined by the relation $pK_a = -\log_{10}K_a$. Values of pK_a are generally of a more convenient magnitude.

Acidity exponent

The above expressions for the various equilibrium constants depend only on the concentrations of the species concerned, which are tacitly assumed to exist in solution independently of one another. This is not always the case, and in exact treatments of these equilibria two modifications are frequently necessary. In the first place, some or all of the reacting species are ions and, because of the electrical forces between them, the law expressing their concentrations at equilibrium is not always valid. Corrections may be applied by multiplying the concentrations by certain factors called activity coefficients, the values of which can be calculated theoretically or derived from other measurements. Furthermore, ions of opposite charge may attract one another so strongly that they no longer exist independently but are partly present as ion pairs, thus altering the forms of the equilibrium equations. For many purposes, however, the simple equations given here are adequate, especially with regard to reactions in aqueous solutions.

Ion product of water

Aqueous solutions. Since aqueous solutions are of particular importance in the laboratory and in the physiology of animals and plants, it is appropriate to consider them separately. The ion product of water, $K_w = [H_3O^+]\ [OH^-]$, has the value 1.0×10^{-14} mole^{2}litre^{-2} at 25° C, but it is strongly temperature dependent, becoming 1.0×10^{-15} at 0° C and 7×10^{-13} at 100° C. In principle the value of K_w can be determined by measuring the electrical conductance of very pure water, in which $[H_3O^+] = [OH^-] = 10^{-7}$ at 25° C, but in practice it is derived from other measurements, for example, measurements of the degree of hydrolysis of salts.

For an uncharged acid, such as acetic, the dissociation constant is given by the following expression:

$$\frac{[CH_3CO_2^-]\,[H_3O^+]}{[CH_3CO_2H]} = K_a.$$

For this example (acetic acid) K_a has the value 1.76×10^{-5} at 25° C. It is often convenient to express the dissociation constant in terms of the degree of dissociation of the acid. This quantity, which is represented in equations by the Greek letter alpha, α, is equal to the fraction of the acid that appears in dissociated form—in this case as the ions $CH_3CO_2^-$ and H_3O^+. If the initial concentration of acid is designated by c, then the concentrations of the ions are each equal to αc, that is $[H_3O^+] = [CH_3CO_2^-] = Ac$, and the concentration of undissociated acid is equal to $c(1 - \alpha)$, or $[CH_3CO_2H] = c(1 - \alpha)$.

Substituting these expressions into the equation giving the value of the dissociation constant gives the following result: $\frac{a^2c}{1-a} = K_a$. From this equation it can be inferred that the degree of dissociation (a) increases with decreasing concentration (c). For small degrees of dissociation ($a < < 1$) the equation becomes $a = \frac{K_a^{\frac{1}{2}}}{c^{\frac{1}{2}}}$ whereas at sufficiently low concentrations ($c < < 1$) a tends to unity ($a \rightarrow 1$).

Discussions exactly analogous to this apply to a number of other acid–base equilibria, for example, (1) the dissociation of ammonia in water, (2) the hydrolysis of ammonium salts, and (3) the hydrolysis of an acetate.

$$(1)\ H_2O + NH_3 \rightleftarrows OH^- + NH_4^+$$

$$(2)\ NH_4^+ + H_2O \rightleftarrows NH_3 + H_3O^+$$

$$(3)\ H_2O + CH_3CO_2^- \rightleftarrows OH^- + CH_3CO_2H$$

Dissociations in water

For reaction (1) α is the degree of dissociation of ammonia, and the dissociation constant is K_b, the basic dissociation constant. In reaction (2), the hydrolysis of an ammonium salt (for example, ammonium chloride), α would be termed the degree of hydrolysis and K the hydrolysis constant. In terms of the general definition of acids and bases, however, K could equally be called the acidity constant for the acid–base pair NH_4^+–NH_3, and this is a more rational way of describing the process. Finally, reaction (3) represents the hydrolysis of an acetate (for example, sodium acetate); the resulting equilibrium constant is termed the hydrolysis constant and can be seen to equal K_w/K_a, where K_w is the ion product of water and K_a the acidity constant for the acid–base pair CH_3CO_2H–$CH_3CO_2^-$; *i.e.*, the dissociation constant of acetic acid. The investigation of equilibria such as this is, in fact, one of the methods for determining the value of K_w (see above).

The equilibria considered so far arise when one component of an acid–base pair is dissolved in water—if necessary, along with an ion, such as Na^+ or Cl^-, having negligible acid-base properties. The direct consequence of this is that the two new species produced (for example, those on the right-hand sides of the equations [1–3] above) have equal concentrations (αc), and hence the previously given equation $K = \frac{a^2c}{1-a}$ is applicable.

Buffered solutions

A solution of a more generally useful type can be obtained by deliberately varying the proportions of acid and base present; such a solution is called a buffered solution or, somewhat more colloquially, a buffer. A buffered solution containing various concentrations of acetic acid and acetate ion, for example, can be prepared by mixing solutions of acetic acid and sodium acetate, by partially neutralizing a solution of acetic acid with sodium hydroxide, or by adding less than one equivalent of a strong acid to a solution of sodium acetate. Similarly, a buffer based on the pair NH_4^+–NH_3 can be prepared by mixing solutions of ammonia and an ammonium salt, by partially neutralizing a solution of ammonia with a strong acid, or by adding less than one equivalent of sodium hydroxide to a solution of an ammonium salt. The hydrogen ion concentration in a buffer solution is, of course, still given by the usual equation, which is conveniently written as

$$[H_3O^+] = K_a\frac{[A]}{[B]}.$$

Hydrogen ion concentration and *p*H

Since hydrogen ion concentrations are usually less than unity, and cover an extremely wide range, it is often convenient to use instead the negative logarithm of the actual concentration, a figure that varies usually only in the range 1–13. This figure is termed the *p*H, and its definition is expressed by the equation $pH = -\log_{10}[H_3O^+]$. For example, in pure water $[H_3O^+] = 1 \times 10^{-7}$, with the result that the $pH = 7.0$. The same term can be applied to alkaline solutions; thus in 0.1 molar sodium hydroxide $[OH^-] = 0.1$, $[H_3O^+] = K_w/[OH^-] = 1 \times 10^{-14}/0.1 = 10^{-13}$, $pH = 13.0$.

Applying the concept of *p*H to buffered solutions gives the following equation: $pH = pK_a - \log_{10}\frac{[A]}{[B]}$ in which $\frac{[A]}{[B]}$, known as the buffer ratio, can be calculated from the way in which the solution is prepared. According to this equation, the *p*H of the buffered solution depends only on the pK_a of the acid and on the buffer ratio. Most particularly it does not depend on the actual concentrations of A and B. Therefore, the *p*H of a buffered solution is little affected by dilution of the solution. It is also insensitive to the addition of acid or alkali, provided that the amounts added are much smaller than both [A] and [B]. This so-called buffering action will be impaired if either [A] or [B] becomes too small; hence, buffer ratios must not deviate too far from unity, and the effective buffering range of a given acid–base system is roughly from $pH = pK_a + 1$ to $pH = pK_a - 1$, corresponding to buffer ratios from 0.1 to 10.

Buffer action

Figure 5 shows the relation between *p*H and composition for a number of commonly used buffer systems. Effective buffer action is confined to the central, steep portion of each curve, where the *p*H is least sensitive to the composition. Figure 5 shows that an acid bearing several acidic hydrogens, such as phosphoric acid, can be used to prepare buffer solutions in several different *p*H ranges. Buffer action plays an important part in controlling the *p*H of many biological fluids; for example, the *p*H of the blood is controlled at about 7.4 by the carbonic acid–bicarbonate system shown in Figure 5. Buffers are widely used to control the *p*H in chemical or biological experiments. For the latter, the system $H_2PO_4^-$–HPO_4^{2-} is particularly useful, being effective in the physiological *p*H range, 6–8.

From N.J. Bjerrum, *Inorganic Chemistry* (1936); Heinemann Educational Books Ltd.

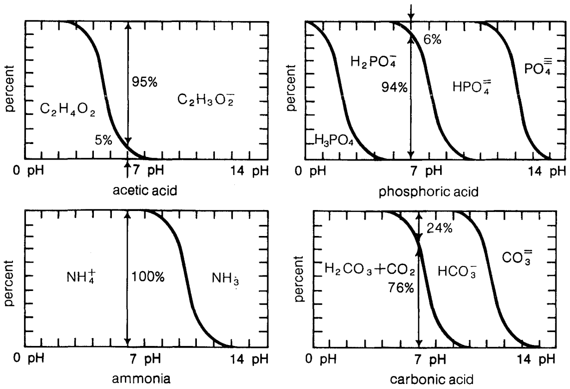


Figure 5: Relation between *p*H and composition for a number of commonly used buffer systems.

The same principles can be applied for the quantitative treatment of systems containing larger numbers of acid–base pairs; for example, in an aqueous solution of ammonium acetate, the following acid–base pairs must be considered: NH_4^+–NH_3, CH_3CO_2H–$CH_3CO_2^-$, H_3O–H_2O, and H_2O–OH^-. The situation is much more complicated in many solutions that are important in industry or in nature, but it is always possible to make a complete prediction of the state of the system in terms of the acidity constants K_a of each acid–base pair (provided, of course, that reactions other than proton transfers do not interfere).

Nonaqueous solvents. Although acid–base properties have been investigated most thoroughly in aqueous solutions, partly because of their practical importance, water is in many respects an abnormal solvent. In particular, it has a higher dielectric constant (a measure of the ability of the medium to reduce the force between two electric charges) than most other liquids, and it is able itself to act either as an acid or as a base. The behaviour of acids and bases in several other solvents will be described briefly here.

Effects of solvent on dissociation

The effect of the solvent on the dissociation of acids or bases depends largely upon the basic or acidic properties of the solvent, respectively. Since many acid–base reactions involve an increase or decrease in the number of ions, they are also influenced by the dielectric constant of the solvent, for a higher dielectric constant favours the formation of ions. Finally, the specific solvation (or close association with the solvent) of particular ions (excluding the solvation of the proton to give SH_2^+, which is already included in the basicity of the solvent) may be important. It is usually not easy to separate these three effects and, in particular, the effects of dielectric constant and solvation merge into one another. These points are illustrated with examples of several of the more important solvents. In this discussion the solvents are classified as amphoteric (both acidic and basic), acidic (in which the acidic properties are much more prominent than the basic), basic (in which the reverse is true), and aprotic (in which both acidic and basic properties are almost entirely absent). Finally, concentrated aqueous acids are mentioned as an example—a particularly important one—of mixed solvents.

Amphoteric solvents. The most important nonaqueous solvents of this class are the lower alcohols methanol and ethanol. They resemble water in their acid–base properties but, because of their lower dielectric constants, facilitate processes producing ions to a much smaller extent. In particular, the ion products of these solvents are much smaller ($K_s = 10^{-17}$ for CH_3OH and 10^{-19} for C_2H_5OH, compared with 10^{-14} for water), and the dissociation constants of molecular acids and bases are uniformly lower than in water by four to five powers of 10. Nitric acid, for example, which is almost completely dissociated in water (K_a about 20), has $K_a = 2.5 \times 10^{-4}$ in methanol. On the other hand, the equilibrium constants of processes such as $NH_4^+ + ROH \rightleftarrows NH_3 + ROH_2^+$ and $CH_3CO_2^- + ROH \rightleftarrows CH_3CO_2H + RO^-$ are similar in all three solvents, since they do not involve any change in the number of ions.

Sulfuric acid as solvent

Acidic solvents. The most important strongly acidic solvent is sulfuric acid, which is able to protonate a wide variety of compounds containing oxygen or nitrogen. Thus, water, alcohols, ethers, ketones, nitro compounds, and sulfones all act as bases in sulfuric acid. This solvent must also possess some basic properties, because its ionic product is high ($[H_3SO_4^+]\,[HSO_4^-] = 1.7 \times 10^{-4}$), but the basicity of the solvent is obscured normally by its very high acidity. For example, carboxylic acids behave as strong bases in sulfuric acid, reacting almost completely according to the equation $RCO_2H + H_2SO_4 \rightarrow RCO_2H_2^+ + HSO_4^-$. Many substances undergo reactions in sulfuric acid that are more complicated than simple proton transfers, often yielding species important because of their chemical reactivity. Thus, some alcohols produce carbonium ions in sulfuric acid; with triphenylcarbinol, for example, the reaction is $(C_6H_5)_3COH + 2H_2SO_4 \rightarrow (C_6H_5)_3\,C^+ + H_3O^+ + 2HSO_4^-$. Nitric acid gives the nitronium ion, NO_2^+, according to the equation $HNO_3 + 2H_2SO_4 \rightarrow NO_2^+ + H_3O^+ + 2HSO_4^-$. This ion frequently is the active agent in the nitration of organic compounds. Hydrogen fluoride has solvent properties resembling those of sulfuric acid, but is less acidic and has negligible basic properties.

Acetic acid is another acidic solvent that has been extensively studied. Because of its low dielectric constant, ions exist in it largely in the form of ion pairs, and more complex associates are frequently formed. For this reason a quantitative interpretation of acid–base equilibria in acetic acid is often difficult, but some general conclusions can be drawn. In particular, it can be seen that all substances more basic in water solution than aniline react completely with acetic acid according to the equation $B + CH_3CO_2H \rightarrow BH^+ + CH_3CO_2^-$. All such bases therefore give solutions with indistinguishable acid–base properties; this is often referred to as a levelling effect of the solvent. The converse is true for acids; for example, the strong mineral acids, nitric, hydrochloric, sulfuric, hydrobromic, and perchloric (HNO_3, HCl, H_2SO_4, HBr, and $HClO_4$) are "levelled" in aqueous solution by complete conversion to the hydronium ion, but in acetic acid they are differentiated as weak acids with strengths in the approximate ratio 1 : 9 : 30 : 160 : 400.

Basic solvents. The only basic solvent that has been investigated in any detail is liquid ammonia, which has the very low ion product $[NH_4^+]\,[NH_2^-] = 10^{-33}$. As might be expected, this solvent has a marked levelling effect upon acids; thus, for example, acetic, benzoic, nitric, and hydrochloric acids all give solutions with identical acidic properties, due to the ion NH_4^+—although, of course, in water they behave very differently.

Aprotic solvents. Strictly aprotic solvents include the hydrocarbons and their halogen derivatives, which undergo no reaction with added acids or bases. Acid–base equilibrium in these solvents can be investigated only when a second acid–base system is added; the usual reaction $A_1 + B_2 \rightleftarrows B_1 + A_2$ then takes place. Most such investigations have employed an indicator as one of the reacting systems, but the results are often difficult to interpret because of association of both ions and molecules in these media of low dielectric constant.

The term aprotic has been extended recently to include solvents that are unable to lose a proton, although they may have weakly basic properties. Some of these aprotic solvents have high dielectric constants (for example, N, N-dimethylformamide, dimethyl sulfoxide, and nitrobenzene) and are good solvents for a variety of substances. They have a powerful differentiating effect on the properties of acids and bases. In particular, basic anions are poorly solvated in these solvents and thus behave as very strong bases; for example, it has been estimated that sodium methoxide dissolved in dimethyl sulfoxide gives a solution 10^9 times as basic as in methanol.

Acidity function

Concentrated aqueous acids. Dilute solutions of strong acids, for example, hydrochloric, sulfuric, and perchloric (HCl, H_2SO_4, $HClO_4$), in water behave essentially as solutions of the ion H_3O^+, and their acidity increases in proportion to their concentration. At concentrations greater than about one molar (that is, containing one mole of acid per litre of solution), however, the acidity, as measured by action on indicators or by catalytic ability, increases much more rapidly than the concentration. For example, a 10 molar solution of any strong acid is about 1,000 times as acidic as a 1 molar solution. This behaviour is undoubtedly largely due to the depletion of water with increasing concentration of acid; the hydronium ion, H_3O^+, is known to have a strong tendency to further hydration, probably mainly to the ion $H_3O^+(H_2O)_3$ (*i.e.*, $H_9O_4^+$), and a decrease in water content increases the proton-donating power of the solution. The acidity of these concentrated solutions is commonly measured by the acidity function, H_0, a quantity measured by the effect of the solvent on a basic indicator I. It is defined by $H_0 + pK_{IH}{}^+ - \log_{10}[IH^+]/[I]$ and becomes equal to the pH in dilute solution. The acidity function H_0 is often found to be independent of the nature of the indicator and to give an approximate measure of the catalytic power of the acid solution. Sulfuric acid–water mixtures ranging from 10 to 100 percent sulfuric acid have H_0 values between −0.3 and −11.1, corresponding to an acidity range of nearly 11 powers of 10.

Lewis acids. Much less information is available about Lewis acid–base equilibria than about ordinary acid–base equilibria, but it is clear that the situation is less simple for the former than for the latter. When a given Lewis acid reacts with a series of similarly constituted bases the equilibrium constants often vary in parallel with the conventional basic strengths. This is the case when a zinc halide, ZnX_2, for example, reacts with a series of amines. In general, however, it is not possible to arrange Lewis acids and bases in a unique order that will predict the extent to which a given pair will react. Thus, although the hydroxide ion (OH^-) is always a much stronger base than ammonia (NH_3) in reactions with proton acids, in reactions with the Lewis acid Ag^+, the complex $Ag(NH_3)_2^+$ is fairly stable, whereas AgOH is completely dissociated. Similarly, for some metal cations complex formation increases in the order fluoride < chloride < bromide < iodide, whereas for other metal cations the order is the reverse of this.

This kind of behaviour has led to a classification of Lewis acids and bases into "hard" and "soft" categories; as a rule, hard acids react preferentially with hard bases and, similarly, soft acids react with soft bases. The terms hard and soft are chosen to suggest that the atomic structures associated with hard acids and bases are rigid and impenetrable, whereas those associated with soft acids and bases are more readily deformable. Hard acids include the proton; sodium, calcium, and aluminum ions; and carbonium ions. The soft acids include cuprous, silver, mercurous, and the halogen cations. Typical soft bases are iodide, thiocyanate, sulfide, and triphenylphosphine; whereas hard bases include hydroxide, fluoride, and many oxyanions. The dividing line between the hard and soft categories is not a sharp one, and its theoretical interpretation is obscure. Nevertheless, a surprising amount of factual information can be coordinated on the basis of preferential reactions of hard acids with hard bases and soft acids with soft bases.

Hard and
soft acids
and bases

The effect of molecular structure. Regularities and trends in the properties of the elements are best understood in terms of the periodic table, an orderly pattern seen when the elements are arranged in order of increasing atomic number. Comparing the hydrides of the various elements in the table reveals appreciable acidity only among those of the elements on the right-hand side of the table, especially the halogen elements—fluorine, chlorine, bromine, and iodine. This generalization is borne out when the elements across the first period of the table are examined in order; as the right-hand side of the table is approached the elements encountered are carbon, nitrogen, oxygen, and fluorine. The hydrides of these elements show increasing acidity. Methane, CH_4, a hydride of carbon, has no detectable acidic properties, and the pK_a decreases sharply in the series ammonia (NH_3), 35; water (H_2O), 16; and hydrogen fluoride (HF), 4. In any given group of the periodic table, the acidity of the hydrides increases as the group is descended. For example, the two groups at the right-hand side of the table include, respectively, and in descending order, the elements oxygen, sulfur, selenium, and tellurium; and fluorine, chlorine, bromine, and iodine. The pK_a's of the hydrides of the first group are as follows: water (H_2O), 16; hydrogen sulfide (H_2S), 7; hydrogen selenide (H_2Se), 4; and hydrogen telluride (H_2Te), 3. Similarly, hydrogen fluoride (HF) is a weak acid, whereas hydrogen chloride (HCl), hydrogen bromide (HBr), and hydrogen iodide (HI) are all completely dissociated (are strong acids) in aqueous solution. These trends are due to variations in bond strength, electronegativity (attractive power of the atomic nucleus for electrons), and ionic solvation energy, of which the first is the most important. When a hydride is able to lose two or more protons, the loss of the second is always more difficult because of the increased negative charge on the base—*e.g.*, $H_2S - HS^-$ (pK 7), $HS^- - S^{2-}$ (pK 15); similarly, $NH_4^+ - NH_3$ (pK 9.5), $NH_3 - NH_2^-$ (pK 35).

A simple rule applies to the strengths of the oxyacids, which can be given the general formula $XO_n(OH)_m$, in which X is any nonmetal. In these compounds, the pK decreases with increasing n, but does not depend significantly upon m. When $n = 0$ (*e.g.*, ClOH, $Si(OH)_4$), pK_a is between 8 and 11; $n = 1$ (*e.g.*, HNO_2, H_2SO_3) gives pK_a 2–4; whereas with $n = 2$ or 3 (*e.g.*, H_2SO_4, $HClO_4$), the acids are completely dissociated in water ($pK_a < 0$).These regularities are probably attributable to the sharing of the negative charge of the anion between $n + 1$ equivalent oxygen atoms; the more extensive the charge spread, the lower is the energy of the anion and hence the stronger the acid.

Strengths
of oxyacids

The most important groups of organic acids are the alcohols (including the phenols) and the carboxylic acids. The simple alcohols are very weak acids (pK 16–19); the phenols are considerably stronger ($pK \sim 10$); and the carboxylic acids stronger still ($pK \sim 5$). The strength of the carboxylic acids is due to the sharing of the negative charge between two equivalent oxygen atoms in the ion RCO_2^-. The most important organic bases are the amines, RNH_2, R_2NH, or R_3N. Most of these are stronger bases than ammonia; *i.e.*, their cations are weaker acids than the ammonium ion.

The effect of substituents on the acid–base properties of organic molecules has been very extensively studied and is one of the main methods of investigating the nature of the electron displacements produced by substitution in these molecules. The simplest classification is into electron-attracting substituents (halogens, carbonyl, nitro, and positively charged groups) and electron-repelling groups (alkyl groups, negatively charged groups). The electron-attracting groups make acids stronger and bases weaker, whereas electron-repelling groups have the opposite effects. There are, however, often more specific electronic effects, especially in aromatic and unsaturated compounds, for which special explanations are needed.

EXPERIMENTAL STUDIES OF ACIDITY AND BASICITY

Dissociation constants in aqueous solution. The classical method for determining the dissociation constant of an acid or a base is to measure the electrical conductivity of solutions of varying concentrations. From these the degree of dissociation (α, see above) can be determined and K_a calculated from the equation:

$$K_a = \frac{\alpha^2 c}{1 - \alpha}.$$

Measurements of
electrical
conductivity

This method is unsuitable for acids with pK less than 2 because α is then close to unity and the value $1 - \alpha$ is therefore subject to error. It also is unsuitable for acids of $pK > 7$, because impurities in the solvent may affect the conductivity or displace the dissociation equilibrium.

It is often preferable to use a more specific method for determining the concentration of one of the species in the scheme $A + H_2O \rightleftarrows B + H_3O^+$. For example, a hydrogen electrode (or more commonly a glass electrode, which responds in the same way) together with a reference electrode, commonly the calomel electrode, serves to measure the actual hydrogen ion concentration, or the pH, of the solution. If E is the electromotive force (in volts) observed by the electrode, the equation giving the pH is as follows:

$$E = E_o + 0.059 \log_{10} [H_3O^+] = E_o - 0.059\, pH.$$

In this equation, the value of E_0 depends on the nature of the reference electrode and is usually obtained by calibration with a solution of known pH. Measurements can be made on aqueous solutions of the acid, in which case $[B] = [H_3O^+]$, but it is better to use a series of buffer solutions with known ratios [A]/[B], since these are less sensitive to the presence of impurities. Such a series is obtained by successive additions of alkali to a solution of the acid (or of a strong acid to a solution of the base) and the procedure is then often termed a pH titration.

If A and B have different optical properties—for example, if they differ in colour or in the absorption of ultraviolet light—this property can be used to measure the ratio [A]/[B], commonly by using an instrument called a spectrophotometer. Since $[H_3O^+]$ must also be known, the commonest procedure is to measure [A]/[B] in a solution made by adding a small quantity of A or B to a standard buffer solution. If A and B do not have convenient optical properties—as is commonly the case—an indicator, that is, an acid–base system that does show a difference in colour in changing from A to B, is used. If a small quantity of indicator A_I–B_I, with acidity constant K_I, is added to a buffer solution A–B, it is easily shown that the following relation holds:

$$K_A = K_I \frac{[B]}{[A]} \cdot \frac{[A_I]}{[B_I]},$$

in which $[A_I]/[B_I]$ is measured spectrophotometrically, and

Table 2: pK_a's of Representative Acids and Bases

	pK_a		pK_a
Inorganic acids		**Carboxylic acids**	
Boric acid	9.1 (20° C)	Formic	3.75 (20° C)
Hypochlorous acid	7.53 (18° C)	Acetic	4.75
Hydrogen sulfide	7.0, 11.9 (18° C)	Chloroacetic	2.85
Carbonic acid	6.4, 10.3	Dichloroacetic	1.48
Phosphoric acid	2.1, 7.2, 12.8	Trichloroacetic	0.70
Sulfurous acid	1.8, 6.9 (18° C)	Oxalic	1.23, 4.19
Nitric acid	−1.6	Malonic	2.83, 5.69
Sulfuric acid	(−3), 1.9	Benzoic	4.19
Hydrogen chloride	(−7)	**Nitrogen bases**	
Perchloric acid	(−8)	Methylamine	10.66
Inorganic bases		Dimethylamine	10.73
Ammonia	9.25	Trimethylamine	9.81
Hydrazine	−0.9, 8.23 (20° C)	Piperidine	11.12
Hydroxylamine	6.03 (20° C)	Aniline	4.63
Alcohols and phenols		Pyridine	5.25
Methanol	15.5	Quinoline	4.90 (20° C)
Trifluoroethanol	12.37	Pyrrole	−0.27
Phenol	9.89		
o-Nitrophenol	7.17		
m-Nitrophenol	8.28		
p-Nitrophenol	7.15		
o-Chlorophenol	8.49		
m-Chlorophenol	8.85		
p-Chlorophenol	9.18		
Picric acid	0.38		

all the other quantities on the right-hand side of the equation are known.

If accurate values of K are required, it is necessary in all the above methods to take into account the effect of interionic forces upon the equation and the quantities measured. This factor can induce a considerable degree of complexity into the problem.

Selected values of acidity constants. Table 2 contains acidity constants for selected substances. These are listed as acids or bases according to the nature of the uncharged species, but in each case the value given is (pK_a) for the acid form (pK_a and pK_b for a conjugate acid–base pair being related by the equation

$$pK_a + pK_b = pK_w = 14.00$$

for aqueous solutions at 25° C). For instances in which several values of pK_a are given, these relate to successive dissociations; *e.g.*, for phosphoric acid they correspond to dissociations of H_3PO_4, $H_2PO_4^-$, and HPO_4^{2-}. All values given refer to aqueous solutions at or near 25° C; parentheses indicate values that have been estimated indirectly, or are uncertain for other reasons. (R.P.B.)

Ion-exchange reactions

As previously noted, ions are atoms, or groups of atoms, that bear a positive or negative electric charge. In pairs or other multiples they make up the substance of many crystalline materials, including table salt. When such an ionic substance is dissolved in water, the ions are freed—to a considerable extent—from the restraints that hold them within the rigid array of the crystal, and they move about in the solution with relative freedom. Certain insoluble materials bearing positive or negative charges on their surfaces react with ionic solutions to remove various ions selectively, replacing them with ions of other kinds. Such processes are called ion-exchange reactions. They are used in a variety of ways to remove ions from solution and to separate ions of various kinds from one another. Such separations are widely utilized in the scientific laboratory to effect purifications and to aid in the analysis of unknown mixtures. Ion-exchange materials such as zeolites are also employed commercially to purify water (among other uses) and medically to serve as artificial kidneys and for other purposes.

EARLY HISTORY

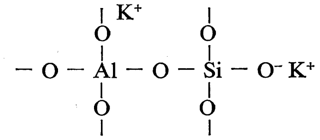
Surprisingly, recognition of ion-exchange processes antedates the great Swedish chemist Svante Arrhenius, who formulated the ionic theory. In 1850, nine years before Arrhenius was born, separate papers appeared in the *Journal of the Royal Agricultural Society of England* by agriculturist Sir H.S.M. Thompson and chemist J.T. Way, describing the phenomenon of ion exchange as it occurs in soils. In his paper, entitled "On the Power of Soils to Absorb Manure," Way addressed himself to the question of how soluble fertilizers like potassium chloride were retained by soils even after heavy rains. Way took a box with a hole in the bottom, filled it with soil, and poured onto the soil a solution of potassium chloride, collecting the liquid that flowed out of the bottom. He then washed the soil with rainwater and analyzed the water he had collected, from both the solution and the rainwater. The water turned out to contain all of the chloride that had been originally added but none of the potassium; the potassium had been replaced by chemically equivalent amounts of magnesium and calcium. Way called the process "base exchange" because of the basic (nonacidic) character of the exchanged elements. That term persisted until after 1940, by which time the process had become universally known as ion exchange.

Base exchange

In modern parlance, the process would be described in the following way: potassium ions enter the soil and displace calcium and magnesium ions. The chloride ions have no part in the operation and pass through unchanged. In terms of a chemical equation, the process is $2K^+ + Ca^{2+}(\text{soil}) \rightleftharpoons Ca^{2+} + 2K^+(\text{soil})$, in which the double arrow indicates that the exchange is reversible. In Way's experiment, the process was pushed to completion (that is, the equilibrium was pushed to the right) because the water trickling through the soil continually came in contact with fresh calcium-loaded soil. As Way also observed, the potassium could be regained by washing the soil with a solution of calcium chloride (which pushed the equilibrium in the opposite direction).

ION-EXCHANGE MATERIALS

Soil is able to bind positive ions (like K^+ and Ca^{2+}) because it contains clay minerals and organic humic acids. Both of those substances are insoluble materials that carry, as a part of their molecular framework, negatively charged ionic groups. In clays, for instance, such groups are the ends of silicon–oxygen chains—either oxygen atoms that carry an extra electron because they are bonded to only one atom instead of the usual two or aluminum atoms bonded to four oxygens instead of the usual three. The following schematic representation shows both kinds of ionic structure as they occur in an almost infinite variety of silicates and aluminosilicates, both natural and artificial.



$$\begin{array}{ccccc} & \mid\ \mathrm{K}^+ & & \mid & \\ & \mathrm{O} & & \mathrm{O} & \\ & \mid & & \mid & \\ -\mathrm{O}- & \mathrm{Al} & -\mathrm{O}- & \mathrm{Si} & -\mathrm{O}^-\mathrm{K}^+ \\ & \mid & & \mid & \\ & \mathrm{O} & & \mathrm{O} & \\ & \mid & & \mid & \end{array}$$

The negative ions are part of the framework; the positive ions, shown here as potassium, are small and can change places with other positive ions if the solid is placed in contact with a solution. The small positive ions must be able to move in and out; they must be located on surfaces or in the interstices of the open lattice structure.

The two requirements for ion exchange—fixed ionic charges on a supporting material and permeability of the material to a solution—are met in a surprisingly large number of materials. The fixed charges may be negative, as in the above example, or they may be positive. The mobile ions must be of opposite charge to the fixed ions. Materials with fixed negative charges (as in Figure 6) exchange positive ions, or cations, and the process is called cation exchange. Those having fixed positive charges correspondingly exchange negative charges, or anions, and are said to undergo anion exchange.

Ion-exchange resins

A big improvement in ion-exchange technology came in 1935, when the first ion-exchange resins were discovered by the English chemists Basil Albert Adams and Eric Leighton Holmes. The resins were chemical relatives of the plastic Bakelite and were made by condensing polyhydric phenols or phenolsulfonic acids with formaldehyde.

In 1944 Gaetano F. D'Alelio patented styrene–divinylbenzene polymers, substances with large, network-like

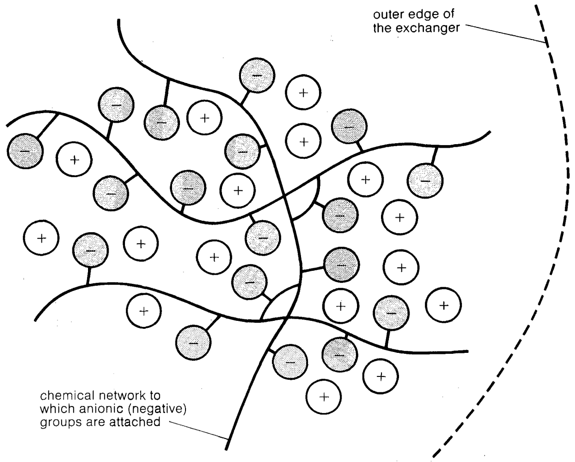
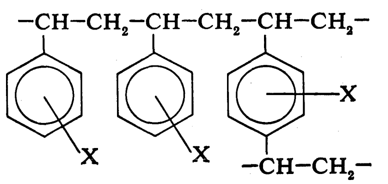


Figure 6: Chemical structure of cation exchanger. The exchangeable ions are marked +. The whole structure is permeated by solvent molecules, usually water (not shown).

molecules, into which ionic groups were introduced by chemical treatment. The structure of these compounds may be represented thus:

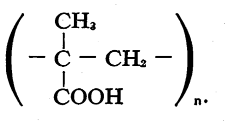


in which X represents the ionic groups, which may occur at various locations on the benzene rings. In the formula as shown, the first two benzene rings come from styrene, whereas the third is from divinylbenzene. Divinylbenzene thus provides cross-linking between the polystyrene chains, joining them into a three-dimensional network that can be tight or loose, depending on the ratio of divinylbenzene to styrene. This ratio can be varied at will; the usual commercial proportion is 8 percent. The ionic groups may be sulfonic acid groups, namely $-SO_3^-H^+$ or quaternary ammonium groups, $-CH_2N^+(CH_3)_3Cl^-$. These two types account for some 90 percent of all ion-exchange resins produced. The hydrogen ions and chloride ions may be replaced by other ions, such as Na^+ (sodium) or OH^- (hydroxide); the hydrogen and hydroxide forms of these resins are very strong acids and bases, respectively.

Styrene and divinylbenzene are liquids and are polymerized as spherical droplets, with the result that the resins have the form of beads that are almost perfect spheres. The beads swell when placed in water, and though they look smooth and impermeable, they are actually very permeable to water and small ions. They may have diameters ranging from a few microns (thousandths of a millimetre) to one to two millimetres. Different sizes are used for different purposes.

Ionic groups other than sulfonic acid and quaternary ammonium salt may be introduced into the resin structure. A useful one is iminodiacetate, $-CH_2N(CH_2COOH)_2$, which forms chelated complexes (structures held together by secondary bonds) with all metals except the alkali metals. The stability of these complexes varies widely from metal to metal. The chelating resins are used in chemical analysis to separate and concentrate trace metals.

Resins carrying the carboxyl group, —COOH—useful in medicine and biochemistry—are based not on polystyrene but on polymethacrylic acid:



$$\left(-\overset{\displaystyle CH_3}{\underset{\displaystyle COOH}{C}}-CH_2-\right)_n.$$

Still another kind of ion exchanger is made from cellulose by introducing various ionic groups into the cellulose molecules. Since the ions are on the surface of the threadlike molecules, instead of being inside molecular frameworks, they are accessible to large ions and molecules. Cellulose-based exchangers are especially useful in biochemistry.

Synthetic inorganic ion exchanges

Synthetic inorganic exchangers have been known since 1903. The first ones were aluminosilicates. About 1955 it was found that phosphates, arsenates, and molybdates of titanium, zirconium, and thorium were good cation exchangers; and many such materials have been prepared, some commercially. They are useful in the nuclear-power industry for they are resistant to radiation and selective to certain radioactive wastes, particularly the long-lived fission product cesium-137. They serve to separate that isotope from other less dangerous fission products.

Another class of inorganic ion exchanger is the molecular sieve. These materials are crystalline aluminosilicates with well-defined structures containing pores of definite sizes that permit only certain ions to enter. When the water is removed from the pores, these substances become selective adsorbents for gas molecules of certain sizes and shapes. They also are powerful catalysts.

The substances termed liquid ion exchangers possibly should be classed as organic solvents, rather than ion exchangers, in spite of their name. The molecules of such substances contain long hydrocarbon chains, which make them insoluble in water, but they also carry ionic groups that attract ions of opposite charge. An example of a liquid ion exchanger is dinonylnaphthalene sulfonic acid, *i.e.,* $(C_9H_{19})_2C_{10}H_5SO_3H$.

ION-EXCHANGE PROCEDURES

Ion-exchange columns

Only rarely are ion exchangers used in stepwise procedures, in which the resin is mixed into a container of solution and then removed for further treatment. Much more frequently the exchanger is packed into a tube or column through which the solution is made to flow. The column arrangement forces the ion-exchange reaction, which is intrinsically reversible, to go to completion in the desired manner. The solution flowing down the column continually meets fresh exchanger, and a reaction that goes half way in the first centimetre of the column may be three-quarters completed in the second centimetre, seven-eighths in the third, and so on. In a short time, the exchangeable ions that entered the column have been adsorbed and become undetectable analytically. When the exchangeable ions do start to emerge from the end of the column, however, the column has become completely saturated with them. It may be restored to its original condition, or regenerated, by passing through it a solution of the ions that it originally contained.

Ion-exchange columns are easy to use, but the theory behind their use is extremely complicated. A column with a solution flowing through it is a nonequilibrium system, and its interpretation must consider not only equilibrium distributions but rates of transfer of material and statistical variations in the paths of liquid flow between the granules that make up the exchanger. There are two chief theories of ion-exchange processes related to the two principal ways in which the columns are employed. In the first procedure, displacement, the column originally contains mobile ions of one kind that are pushed down the column by the steady flow of a solution of ions of a second kind. The theory for this procedure deals with the rate at which the "front" (the boundary between the different classes of ions) advances and its concentration profile (whether the front stays sharp as it moves down the column or whether it becomes progressively more diffuse). In the second process, elution, a thin layer of ions is introduced at the top of a column already saturated with ions of a second kind; it then is washed down the column with the same kind of ions that saturated the column at the start. In elution, the theory must account for the rate of movement of the narrow band of ions and its spread as it proceeds down the column.

When a mixture of two kinds of ions that are held by the exchanger with differing strengths is introduced at the top of a column, the mixture of ions separates as it moves down the column, with the result that the original single band of ions is resolved into two separate bands. This process is called ion-exchange chromatography. Ion-exchange chromatography is an important tool in chemical analysis because it permits separation of materials that are very difficult to separate by other means. It can be applied to organic and inorganic ions and even to substances that are not ionic. It is often used to separate mixtures of many components.

Ion-exchange columns are made in all sizes, from the large tanks used to soften the water supply of great cities to the tiny columns holding less than a cubic centimetre of resin that are used for recovering short-lived radioactive elements in the laboratory. (The element mendelevium was discovered by the isolation of a few atoms on an ion-exchange column.) Care is needed in preparing columns for the laboratory. Dry resin must be stirred with water to let it swell before it is poured into the column. Air bubbles must not be allowed to form in the resin bed for they interfere with the liquid flow. It is desirable to backwash the column—that is, to pass liquid upward to expand the resin bed—in order to release air bubbles and to segregate the resin particles according to size before the column is used. The aim of the preparation procedure is to assure even packing and even flow. For difficult chromatographic separations, resins having uniform, very small particles are used to facilitate mass transfer and give sharper bands. As fine particles offer much resistance to flow, solutions must be forced through under pressure. One procedure is to use long, narrow columns of stainless steel (like those used in a related process called gas chromatography) and to hasten mass transfer by using, instead of resin beads, glass spheres coated with ion-exchange resin.

Preparation of columns

Ion exchangers, especially inorganic and cellulose-based exchangers, are used in thin-layer chromatograpy. Chromatographic paper for this purpose is manufactured from finely ground resins and cellulose fibres. One use for this procedure is to filter small traces of metal ions from large volumes of solution.

Ion-exchange resins also may be fabricated into thin sheets, although it is not easy to make a sheet of ion exchanger that is strong and flexible and at the same time permeable; development of ion-exchange membranes has been slow for this reason. Ion-exchange membranes are used, however, to separate the electrodes of fuel cells and to remove salts from water by the physical processes termed reverse osmosis and electrodialysis. The former is a kind of filtration process—water is squeezed through the membrane under pressure while the dissolved salts are left behind. The reaction can be carried out, for example, by placing a membrane of cation-exchange resin loaded with sodium ions in contact with a dilute solution of sodium chloride. Because of the characteristics of the ion-exchange process, neither the sodium nor the chloride ions can enter the membrane. Water molecules can penetrate the membrane, however; and because of the pressure exerted on the system, they do so, crossing to the other side. The result is the removal of salt from the water without distillation (the usual desalting process).

Ion-exchange membranes

Electrodialysis is a process somewhat similar to reverse osmosis. Ions are able to enter an ion-exchange membrane if they are simultaneously removed from it at the other side; the effect is the same as passing an electric current through the membrane. In practice electrodialysis is carried out by placing a cation-exchange membrane on one side of the solution to be desalted and an anion-exchange membrane on the other and then passing an electric current through the system. The result is that positive ions pass through the membrane on one side and negative ions pass through the membrane on the other. Pure water is left eventually in the area between the membranes.

APPLICATIONS OF ION EXCHANGE

In the laboratory. Ion exchange is used for both analytical and preparative purposes in the laboratory, the analytical uses being the more common. An important use of ion-exchange chromatography is in the routine analysis of amino-acid mixtures. Columns of cation-exchange resin are used, and the solutions are maintained sufficiently acid that the amino acids are at least partly in their cationic forms. The 20 principal amino acids from blood serum or from the hydrolysis of proteins are separated in a few hours, and their concentrations are determined automatically by light-absorption methods. Such analysis is used in clinical diagnosis.

In a less routine and highly important application of ion-exchange chromatography, the products of hydrolysis of nucleic acids are analyzed. In this way, information is gained about the structure of these molecules and how it relates to their biological function as carriers of hereditary information. Cation-exchange resins are used for this purpose as well. Because of their use in analyzing the structures of complex biological materials, ion-exchange chromatographic procedures have been of great importance in the development of modern molecular biology—the explanation of biological processes in terms of the interactions of molecules.

Inorganic ions also can be separated by ion-exchange chromatography. The lanthanides, or rare-earth elements, are separated on columns of cation-exchange resin. Solutions of citrates, lactates, or other salts whose anions form negatively charged complexes with the lanthanide ions are used to wash the ions from the column. The metal ions themselves are held by the resin; the complexes are not. Those ions that form more stable complexes do not adhere to the resin and therefore move off the column more quickly than the ions that do not form complexes (or complex only weakly). Cation exchange in general is not a selective process, but the above process, termed differential complex formation, renders it more so. In lanthanide separations the exchanger is like an undiscriminating sponge that simply holds the metal ions, whereas the real separation of the various metals is accomplished by the weakness or strength of the complexes formed.

Separation of inorganic ions

Anion exchange in hydrochloric acid is an effective way to separate metal ions. Most metals form negatively charged chloride complexes that can be held by anion-exchange resins carrying quaternary ammonium groups (NR_4^+). These complexes differ greatly in their stabilities in solution and in their affinities for the resin. The distribution of metal ions between the solution and the resin depends on the hydrochloric acid concentration and the identity of the metal ion. Impressive separations of metal ions can be achieved by manipulating the hydrochloric acid concentration.

Ion-exchange separations of this kind are widely used; they can be modified by using mixed solvents, like acetone–water, and great selectivity is possible. In the process called "activation analysis," an unknown sample to be analyzed is bombarded with neutrons, and the radioactive elements thus formed are separated by anion-exchange procedures. Such analysis is especially valuable in separating minor metallic constituents from samples containing large amounts of other substances. The technique has been used to analyze lunar rocks.

Chelating resins are used to collect trace metals from seawater. Further, a copper-loaded chelating resin also adsorbs, by coordination, traces of amino acids from seawater. Miscellaneous analytical uses of ion exchange include the dissolving of sparingly soluble salts like calcium sulfate, the determination of total dissolved salts in natural waters (by passing them through hydrogen-loaded, cation-exchange resins and titrating the acid formed), and the identification of minute traces of ions (by absorbing them onto a single resin bead along with a colour-producing reagent).

Preparative uses of ion exchange in the laboratory are not many, but on occasion unusual acids, such as hydroferrocyanic acid, or unusual bases, like cesium hydroxide, are prepared from their salts by passing solutions of the salts through the appropriate resins. Resins also are used to purify acids or bases that contain nonionic contaminants, and to remove ionic contaminants from solvents.

In industry and medicine. Ion exchange finds its major industrial application in the treatment of water. Hard water—caused by the presence of calcium and magnesium ions, which form insoluble precipitates with soaps—is softened by exchanging its calcium and magnesium ions with sodium ions. To accomplish this, the hard water is

Water softening by ion exchange

passed through a column of cation exchanger containing sodium ions. After the column has been in use for some time, calcium and magnesium begin to appear in the water leaving the column. Then the column must be regenerated by passing a concentrated solution of common salt slowly through the column; the excess sodium ions displace the ions that produce the hardness so that, after flushing with water, the bed of exchanger is ready to be used again. At first, the exchangers used for this purpose were natural aluminosilicates; but later, synthetic resins came to be used instead.

For special purposes, such as use in the laboratory, water is deionized—that is, freed entirely from dissolved ions of all kinds. This is accomplished by passing the water through two resin beds in separate columns. The first bed contains a cation-exchange resin bearing hydrogen ions and converts the dissolved salts to their free acids. The second contains an anion-exchange resin loaded with hydroxyl ions; it neutralizes the acids, holding back their anions, and leaves nothing in the water but nonionic impurities. The beds are regenerated by strong acid and strong alkali, respectively. An alternative procedure, "mixed-bed" deionization, uses only one column containing the two resins mixed. Since the resins must be separated for regeneration, however, mixed beds are used chiefly in disposable cartridges for small laboratory units.

Resins used for water treatment should last for many years, but their life may be shortened either by accumulation of colloidal matter (prevented by adding activated carbon filters) or by oxidation caused by the dissolved chlorine in the water. Quaternary-base anion-exchange resins carrying hydroxyl ions also deteriorate; they decompose slowly to give tertiary amine polymers and methanol.

In an even older use of ion exchange, salts are removed from sugar juices to raise the yield of crystallized sugar. Deionization also can improve the flavour and storage time of pineapple juice and wine. In these and other beverage applications, ion exchange removes traces of heavy metals, which not only taste bad but also catalyze oxidation.

In hydrometallurgy, the treatment of ores with water solutions, ion exchange helps to recover valuable metals like copper, silver, and gold from waste waters. Uranium can be recovered from low-grade ores by leaching with dilute sulfuric acid—oxidizing if necessary to convert uranium(IV) to uranium(VI)—and then absorbing the negatively charged uranium sulfate complex ions on a quaternary-base anion-exchange resin. This highly selective absorption process thereby separates the uranium from iron and other metals. The uranium is later removed from the resin with dilute nitric acid.

On an industrial scale, cation exchange separates rare-earth elements by means of a displacement technique in which each element displaces elements bound less strongly than it is as it proceeds down the column. The elements emerge (the one with the weakest bond first) one after the other in high purity.

Ion exchangers can function as catalysts. Strong-acid cation-exchange resins loaded with hydrogen ions catalyze certain chemical reactions carried out in the liquid phase, such as hydrolysis and esterification (ester formation). The advantage of the resin over hydrochloric acid as a catalyst in these reactions is that it is present as a separate phase that does not contaminate the product. In addition, the ion-exchange process lends itself to continuous-flow techniques. Gas-phase reactions catalyzed by metal ions, like the cracking of petroleum fractions to produce gasoline, also can be catalyzed by metal-loaded inorganic exchangers, the molecular sieves being particularly suitable for this purpose since their open crystalline structure makes every metal ion accessible.

Ion exchange in medicine

Ion-exchange resins have a limited use in medicine. Carboxylic resins containing hydrogen or ammonium ions, taken by mouth, remove sodium ions from the gastrointestinal tract and control edema; other resins are consumed to lower acidity in the stomach and hence to soothe stomach ulcers. Interest in these treatments has declined, however, because of the resins' undersirable side effects. Resins also are incorporated into artificial kidneys outside the body to remove ammonium and potassium ions from the blood. The most important medical applications of ion exchange, however, have been made in clinical analysis procedures that depend on ion-exchange chromatography.

EQUILIBRIA AND KINETICS

Ion-exchange equilibria. The reversibility of ion-exchange reactions greatly affects the behaviour of ion-exchange systems. Typical ion-exchange reactions can be written as follows:

$$A^+ + B^+Res^- \rightleftharpoons B^+ + A^+Res^-$$
$$C^{2+} + 2B^+Res^- \rightleftharpoons 2B^+ + C^{2+}(Res^-)_2,$$

in which Res^- stands for an ion fixed in the resin or other type of exchanger and A^+ and B^+ are univalent cations (monopositive ions) while C^{2+} is a divalent cation.

As is generally true of reversible reactions, equations can be written describing the relative concentrations (amount of material per unit volume) of the various species in equilibrium—that is, when the rate of the forward reaction is equalled by that of the reverse reaction. For the ion-exchange processes indicated above, the following equations demonstrate the relations of the materials present under the conditions of equilibrium:

$$K_1 = \frac{[B^+]\,[A^+\,Res^-]}{[A^+]\,[B^+\,Res^-]};\; K_2 = \frac{[B^+]^2\,[C^{2+}\,Res_2^-]}{[C^{2+}]\,[B^+\,Res^-]^2}.$$

In these equations, the constant K_1 is a pure number without units (such as feet per second) because the units on the right side of the equation cancel. The numerical value of K_2, however, as well as its units, depends on the units chosen to express the concentrations on the right side of the equation. Since the constant K_1 or K_2 is the most useful simple description of the equilibrium conditions of an ion-exchange reaction (and knowledge of its value permits calculation of the concentrations of the various substances in equilibrium under specified conditions), determinations of K values are fundamental steps in the study of ion-exchange reactions. For concentrations in the resin, the proportions of mobile ions to the resin framework can be used, or one can refer the amounts of mobile ions to the weight of water that the resin contains. More constant values of K_2 can be obtained by reference to the resin rather than to the water, which is fortunate, because the water content of a swollen resin depends on the mobile ions it contains and it is hard to measure. As experimentally determined, however, the values of K_1 and K_2 are not constant but vary with the proportions of exchangeable ions in the resin. This result leads to the unsurprising conclusion that the interior of the resin is not an ideal solution.

The distribution of ions of unequal charge, like B^+ and C^{2+} above, depends on the total concentration of the solution. The more dilute the solution, the greater the tendency for the ions of higher charge to accumulate in the exchanger. One doubly-charged ion entering the exchanger sends two singly-charged ions back into the solution, and the more dilute the solution, the more likely is this replacement to happen. This effect, called "electroselectivity," is used in water softening. Calcium ions are taken up from hard water, which is a dilute solution, whereas they are removed from the resin by regenerating with a solution of sodium chloride that is concentrated.

Cation exchange

Many studies have been made of ion-exchange equilibria and the factors that influence the values of K. Different ions are held by exchangers with different strengths. As yet it is impossible to predict a priori the magnitude of K, but one can make certain generalizations, which are different for cation and anion exchange. For the alkali and alkaline-earth metal ions, the strength of binding varies inversely as the ionic hydration. Thus, lithium ions, the most strongly hydrated of the alkali metal ions, are the most weakly held by resins, followed in order by sodium, potassium, rubidium, and finally cesium, which forms the strongest bond with the resins. In the alkaline earths the increasing order runs from beryllium to magnesium, calcium, strontium, barium, and to radium, which is the most strongly held.

These sequences are characteristic of resins whose functional group is the sulfonate ion. Resins bearing carboxylate ions, or with fully ionized phosphonate ions, exhibit different sequences. The electrostatic field strength of the fixed ion on the resin determines the order of separation. When the charge on the fixed ion is small and spread over a large area, as in the sulfonate ion, $-SO_3^-$ the field strength is weak and the mobile ions keep their primary hydration shell—that is, the water molecules they hold by direct coordination. The more strongly hydrated ions migrate to where there is more water—that is, out of the resin and into the surrounding solution.

When the ionic charge is concentrated, however, as it is around the terminal oxygens of silicate networks, the mobile cations are attracted so strongly that the primary hydration shell may be squeezed out. The fixed and mobile ions then come into direct contact. The smaller the mobile ion the closer it gets to the fixed ion and the greater the force that holds it. When the ionic charge is extremely concentrated, the alkali-metal sequence is completely reversed: lithium is held most strongly and cesium the least strongly. Intermediate sequences are possible. In this way the selectivity orders in glass electrodes and biological membranes, both of which function by competitive surface adsorption of positive ions, can be explained. This theory is of little help, however, in explaining selectivity orders of heavy-metal ions; in this case, other factors, as yet unknown, seem to be at work.

Anion exchange

The selectivity sequence for halide ions in resins with quaternary-ammonium fixed ions is fluoride, chloride, bromide, and iodide, with fluoride being held the most weakly. This resembles the cation selectivity order, in which the smallest ion also is held most weakly. On the other hand, the differences between the various ions in degree of attachment to the resin is much greater in the series of halide anions than in the series of alkali-metal cations. Iodide is held a hundred times as strongly as fluoride in an 8 percent cross-linked quaternary-ammonium resin; whereas in an 8 percent cross-linked sulfonic resin, cesium is held four times as strongly as lithium.

It is significant that anions are larger than cations (in their crystal-lattice radii) and that they interact with water in a different way. Instead of attracting dipolar water molecules around them, anions tend to break up the hydrogen-bonded structure of liquid water, with the result that the bigger they are, the more difficult it is for them to enter the water. Large ions are thus driven out of the water and into the resin—a phenomenon of great practical value in achieving separations. Perchlorate ions, ClO_4^-, are held 10 times as strongly as iodide ions. This effect extends to complex ions such as the chlorides of iron and gold, $FeCl_4^-$ and $AuCl_4^-$, which are held very strongly by quaternary ammonium-type anion-exchange resins. The effect of large size may sometimes be offset by an increased ionic charge, which tends to orient the water molecules and stabilize the dissolved ion. Thus, the higher charged zinc complex $ZnCl_4^{2-}$, with its double negative charge, is more weakly bound by the resin than the iron complex $FeCl_4^-$, with its single negative charge.

Very large ions, of course, cannot enter the resin network. With ions of moderate size, the entry of the ion into the network may become a limiting factor, and it becomes important to distinguish between unfavourable equilibrium, on the one hand, and slow exchange, on the other. The equilibrium absorption may be large, but it may take a long time to reach it because the ion has difficulty entering the resin.

Ion-exchange kinetics. Generally, ion exchange is fast. No electron-pair bonds need be broken, and the rate of the process is limited only by the rate at which ions can diffuse in and out of the exchanger structure. The openness of the resin structure, however, depends on the degree of its swelling and on its water content. When both swelling and water content are small, diffusion rates are correspondingly slow. This is true of carboxylic-acid resins and of chelating resins in their acid (hydrogen-ion) form, for these acids are weakly ionized. It is also true of metal-loaded chelating resins. These types of exchangers require ample time for reaction; thus, it is advisable to use a resin with fine particles. High temperature also hastens diffusion. Nonetheless, the general rapidity and efficiency of their actions have brought widespread acceptance and use of ion exchangers. (H.F.W.)

Oxidation–reduction reactions

In the classification of chemical reactions, the terms oxidation and reduction cover a large and diverse body of processes. Many oxidation–reduction, or redox, reactions are as common and familiar as fire, the rusting and dissolution of metals, the browning of fruit, and respiration and photosynthesis—basic life functions.

GENERAL CONSIDERATIONS

Characteristics of redox processes

Major classifications. Most oxidation–reduction (redox) processes involve the transfer of oxygen atoms, hydrogen atoms, or electrons, with all three processes sharing two important characteristics: (1) they are coupled—*i.e.,* in any oxidation reaction a reciprocal reduction occurs; (2) they involve a characteristic net chemical change—*i.e.,* an atom or electron goes from one unit of matter to another. Both reciprocity and net change are illustrated below in examples of the three most common types of oxidation–reduction reactions.

Oxygen-atom transfer. Carbon reacts with mercury(II) oxide (a compound in which mercury has a bonding capacity expressed as +2; see below *Oxidation-state change*) to produce carbon dioxide and mercury metal. This reaction can be written in equation form:

$$\underset{\text{carbon}}{C} + \underset{\substack{\text{mercury(II)}\\\text{oxide}}}{2HgO} \longrightarrow \underset{\substack{\text{carbon}\\\text{dioxide}}}{CO_2} + \underset{\substack{\text{mercury}\\\text{metal}}}{Hg}.$$

Carbon, receiving oxygen, is oxidized; mercury(II) oxide, losing oxygen, undergoes the complementary reduction; and the net change is the transfer of two oxygen atoms from mercury(II) oxide units to a carbon atom.

Hydrogen-atom transfer. Hydrogen atoms are transferred from hydrazine, a compound of nitrogen and hydrogen, to oxygen in the following reaction:

$$\underset{\text{hydrazine}}{N_2H_4} + \underset{\text{oxygen}}{O_2} \longrightarrow \underset{\text{nitrogen}}{N_2} + \underset{\text{water}}{2H_2O}.$$

Hydrazine, losing hydrogen, is oxidized to molecular nitrogen, while oxygen, gaining hydrogen, is reduced to water.

Electron transfer. Zinc metal and copper(II) ion react in water solution, producing copper metal and an aqueous (denoted by aq) zinc ion according to the equation:

$$\underset{\substack{\text{zinc}\\\text{metal}}}{Zn} + \underset{\substack{\text{aqueous}\\\text{copper(II)}\\\text{ion}}}{Cu^{2+}(aq)} \longrightarrow \underset{\substack{\text{aqueous}\\\text{zinc(II)}\\\text{ion}}}{Zn^{2+}(aq)} + \underset{\substack{\text{copper}\\\text{metal}}}{Cu}.$$

With the transfer of two of its electrons, the zinc metal is oxidized, becoming an aqueous zinc ion, while the copper(II) ion, gaining electrons, is reduced to copper metal. Net change is the transfer of two electrons, lost by zinc and acquired by copper.

Because of their complementary nature, the oxidation and reduction processes together are referred to as redox reactions. The reactant that brings about the oxidation is called the oxidizing agent, and that reagent is itself reduced by the reducing agent. In the examples given, mercury(II) oxide, oxygen, and the copper(II) ion are oxidizing agents, and carbon, hydrazine, and zinc are the reducing agents.

General theory. *Stoichiometric basis.* Describing the redox processes as above conveys no information about the mechanism by which change takes place. A complete description of the net chemical change for a process is known as the stoichiometry of the reaction, which provides the characteristic combining proportions of elements and compounds. Reactions are classified as redox and nonredox on the basis of stoichiometry; oxygen-atom, hydrogen-atom, and electron transfer are stoichiometric categories.

Oxidation-state change. Comprehensive definitions of oxidation and reduction have been made possible by modern molecular structure theory. Every atom consists of a positive nucleus, surrounded by negative electrons, which determine the bonding characteristics of each element. In forming chemical bonds, atoms donate, acquire, or share electrons. This makes it possible to assign every atom an oxidation number, which specifies the number of its electrons that can be involved in forming bonds with other atoms. From the particular atoms in a molecule and their known bonding capacities, the bonding pattern within a molecule is determined, and each atom is regarded as being in a specific oxidation state, expressed by an oxidation number.

Redox processes defined

Redox processes are defined as reactions accompanied by oxidation-state changes: an increase in an atom's oxidation number corresponds to an oxidation; a decrease, to a reduction. In this generalized theory, three examples of ways in which oxidation-state changes can occur are by oxygen-atom (gain, oxidation; loss, reduction), hydrogen-atom (loss, oxidation; gain, reduction), and electron (loss, oxidation; gain, reduction) transfer. The oxidation-state change definition is usually compatible with the above rules for applying the oxygen-atom-transfer and hydrogen-atom-transfer criteria and always compatible with the electron-transfer criterion when it is applicable. The oxidation state of any atom is indicated by a roman numeral following the name or symbol for the element. Thus: iron(III), or Fe(III), means iron in an oxidation state of +3. The uncombined Fe(III) ion is simply Fe^{3+}.

Historical origins of the redox concept. Of the chemical processes now regarded as redox reactions, combustion was the earliest focus of philosophical and scientific attention. The Greek scientific philosopher Empedocles listed fire as one of the four elements of matter. In more modern times the phlogiston theory enjoyed scientific popularity. This theory was first articulated in 1697 by G.E. Stahl of Germany. As noted earlier, it asserted that matter releases an elementary constituent, phlogiston, during combustion. Thus, the burning of charcoal was interpreted as the loss of phlogiston from carbon to the air. The theory was also applied to processes other than combustion; in the recovery of a metal from its oxide by heating with charcoal, for example, phlogiston was regarded as being transferred from carbon to the oxide.

Phlogiston saturation was believed to be responsible for the limited ability of air in a closed container to support combustion. A notable consequence of the phlogiston theory was the notion that an oxide of a metal, such as mercury(II) oxide (HgO), was a chemically simpler substance than the metal itself: the metal could be obtained from the oxide only by the addition of phlogiston. The phlogiston theory, however, could provide no acceptable explanation of the gain in weight when an oxide is formed from a metal.

The work of Priestley and Lavoisier

Combustion and oxide formation. Late in the 18th century, the interrelated work of Joseph Priestley and Antoine-Laurent Lavoisier led to the overthrow of the phlogiston theory. Lavoisier saw Priestley's discovery of oxygen in 1774 as the key to the weight gains known to accompany the burning of sulfur and phosphorus and the calcination of metals (oxide formation). In his *Traité élémentaire de chimie*, he clearly established that combustion consists of a chemical combination between oxygen from the atmosphere and combustible matter (see below *Combustion and flame*). By the end of the century, his ideas were widely accepted and had been successfully applied to the more complex processes of respiration and photosynthesis. Reactions in which oxygen was consumed were classified as oxidations, while those in which oxygen was lost were termed reductions.

Electrochemical reactions. During the 19th century, the evolving field of electrochemistry led to a broadened view of oxidation. It was possible, for instance, to produce the ferric, or iron(III), ion from the ferrous, or iron(II), ion at the anode (positive electrode, where electrons are absorbed from solution) of an electrochemical cell (a device in which chemical energy is converted to electrical energy), according to the equation:

$$\underset{\text{iron(II) ion}}{\mathrm{Fe^{2+}(aq)}} \longrightarrow \underset{\text{iron(III) ion}}{\mathrm{Fe^{3+}(aq)}} + \underset{\text{electron}}{e^-} \;.$$

Molecular oxygen could effect a similar transformation, according to the equation:

$$\underset{\text{iron(II) ion}}{\mathrm{4Fe^{2+}(aq)}} + \underset{\text{oxygen}}{\mathrm{O_2}} + \underset{\text{hydrogen ion}}{\mathrm{4H^+}} \longrightarrow \underset{\text{iron(III) ion}}{\mathrm{4Fe^{3+}}(aq)} + \underset{\text{water}}{\mathrm{2H_2O}} \,.$$

The similarity of the two processes led to a precursor of the electron-transfer explanation for redox reactions. After the discovery of the electron, the conviction that oxidation and reduction are accomplished through electron loss and gain became firmly entrenched. Thus, early in the 20th century chemists tended to attribute all redox reactions to the transfer of electrons. Later work on chemical bonding, however, demonstrated the incorrectness of that description. An electronegativity scale (listing of elements in descending order of their tendency to attract and hold bonding electrons) provided a firm basis for the oxidation-state assignments on which oxidation–reduction definitions have become based.

Examples of oxidation–reduction reactions. Molecular oxygen is a conspicuously important oxidizing agent. It will directly oxidize all but a few of the metals and most of the nonmetals as well. Often these direct oxidations lead to normal oxides such as those of lithium (Li), zinc (Zn), phosphorus (P), and sulfur (S).

$$\mathrm{4Li + O_2 \longrightarrow 2Li_2O}\ \text{(lithium oxide)}$$

$$\mathrm{2Zn + O_2 \longrightarrow 2ZnO}\ \text{(zinc oxide)}$$

$$\mathrm{4P + 5O_2 \longrightarrow P_4O_{10}}\ \text{(phosphorus pentoxide)}$$

$$\mathrm{S + O_2 \longrightarrow SO_2}\ \text{(sulfur dioxide)}$$

Oxidation of organic foodstuffs

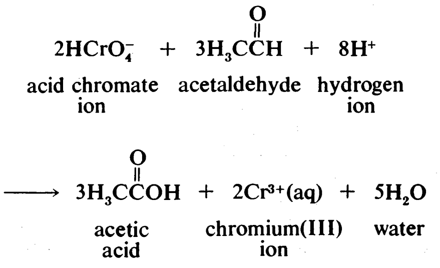
Organic foodstuffs are oxidized to carbon dioxide and water in respiration. The reaction stoichiometry can be illustrated for glucose, a simple sugar:

$$\underset{\text{glucose}}{\mathrm{C_6H_{12}O_6}} + \underset{\text{oxygen}}{\mathrm{6O_2}} \longrightarrow \underset{\text{carbon dioxide}}{\mathrm{6CO_2}} + \underset{\text{water}}{\mathrm{6H_2O}} \,.$$

Although the oxygen–glucose reaction is slow at ambient temperatures outside the living cell, it proceeds quickly under the influence of enzymatic catalysis within the body. Essentially all organic compounds react with oxygen under appropriate conditions, but the reaction rates at ordinary temperatures and pressures vary greatly.

Many other oxidizing agents serve as oxygen-atom sources. Hydrogen peroxide (H_2O_2), acid chromate ion ($HCrO_4^-$), and hypochlorous acid (HClO) are reagents often used in oxygen-atom-transfer reactions, for example in the following reactions:

$$\underset{\text{hydrogen peroxide}}{\mathrm{H_2O_2}} + \underset{\text{chloride ion}}{\mathrm{Cl^-}} \longrightarrow \underset{\text{hypochlorite ion}}{\mathrm{OCl^-}} + \underset{\text{water}}{\mathrm{H_2O}}$$



$$\underset{\text{acid chromate ion}}{\mathrm{2HCrO_4^-}} + \underset{\text{acetaldehyde}}{\mathrm{3H_3C\overset{\overset{\displaystyle O}{\|}}{C}H}} + \underset{\text{hydrogen ion}}{\mathrm{8H^+}}$$

$$\longrightarrow \underset{\text{acetic acid}}{\mathrm{3H_3C\overset{\overset{\displaystyle O}{\|}}{C}OH}} + \underset{\text{chromium(III) ion}}{\mathrm{2Cr^{3+}(aq)}} + \underset{\text{water}}{\mathrm{5H_2O}}$$

$$\underset{\text{hypochlorous acid}}{\mathrm{5HClO}} + \underset{\text{iodine}}{\mathrm{I_2}} + \underset{\text{water}}{\mathrm{H_2O}} \longrightarrow \underset{\text{hydrogen chloride}}{\mathrm{5HCl}} + \underset{\text{iodic acid}}{\mathrm{2HIO_3}} \,.$$

In the simplest hydrogen-atom transfers, molecular hydrogen serves as the hydrogen-atom source. The hydrogenations of ethylene and of molecular nitrogen are illustrative in the following equations:

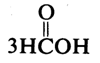
$$\underset{\text{ethylene}}{H_2C{=}CH_2} + \underset{\text{hydrogen}}{H_2} \longrightarrow \underset{\text{ethane}}{H_3CCH_3}$$

$$\underset{\text{nitrogen}}{N_2} + \underset{\text{hydrogen}}{3H_2} \longrightarrow \underset{\text{ammonia}}{2NH_3}\,.$$

Reactions of molecular hydrogen are characteristically slow at ordinary temperatures. The hydrogenation of molecular nitrogen and of olefins such as ethylene (an olefin is an unsaturated hydrocarbon compound; it has at least two adjacent carbon atoms joined by a double bond to which other atoms or groups of atoms can be joined directly) is a process of extraordinary commercial importance and requires catalysts to occur at useful rates.

Hydrogen-atom transfer in organic oxidation

Hydrogen-atom transfer from an organic molecule to a suitable acceptor is a common mode of organic oxidation. The oxidation of formic acid by permanganate and that of ethanol by acid chromate share stoichiometry that features hydrogen-atom loss by the organic species, as shown in the following equations:



$$\underset{\text{formic acid}}{3H\overset{\overset{\displaystyle O}{\|}}{C}OH} + \underset{\text{permanganate ion}}{2MnO_4^-} + \underset{\text{hydrogen ion}}{2H^+}$$

$$\longrightarrow \underset{\text{carbon dioxide}}{3CO_2} + \underset{\text{manganese dioxide}}{2MnO_2} + \underset{\text{water}}{4H_2O}$$

$$\underset{\text{ethanol}}{3CH_3CH_2OH} + \underset{\text{acid chromate ion}}{2HCrO_4^-} + \underset{\text{hydrogen ion}}{8H^+}$$

$$\longrightarrow \underset{\text{water}}{8H_2O} + \underset{\text{acetaldehyde}}{3CH_3CHO} + \underset{\text{aqueous chromium(III) ion}}{2Cr^{3+}(aq)}$$

The oxidizing agents permanganate and acid chromate, typical of many hydrogen-atom acceptors, undergo complicated changes rather than simple hydrogen-atom addition.

Electron-transfer stoichiometry is usually associated with metal ions in aqueous solution, as shown in the following equations:

$$\underset{\text{iron(II) ion}}{Fe^{2+}(aq)} + \underset{\text{cobalt(III) ion}}{Co^{3+}(aq)} \longrightarrow \underset{\text{iron(III) ion}}{Fe^{3+}(aq)} + \underset{\text{cobalt(II) ion}}{Co^{2+}(aq)}$$

$$\underset{\text{thallium(I) ion}}{Tl^+(aq)} + \underset{\text{cerium(IV) ion}}{2Ce^{4+}(aq)} \longrightarrow \underset{\text{thallium(III) ion}}{Tl^{3+}(aq)} + \underset{\text{cerium(III) ion}}{2Ce^{3+}(aq)}$$

Many positively charged metal ions have been shown to be bonded to water molecules, so that their electron-transfer reaction occurs between rather complex molecular groups. The iron ion formulas above, for example, are more properly written as $[Fe(H_2O)_6]^{2+}$ and $[Fe(H_2O)_6]^{3+}$ to reflect the presence of six water molecules bonded to the metal ion. Simple electron transfer between free ions is known only in the gas phase, as in this argon–sodium reaction:

$$\underset{\text{argon ion}}{Ar^+(g)} + \underset{\text{sodium}}{Na(g)} \longrightarrow \underset{\text{sodium ion}}{Na^+(g)} + \underset{\text{argon}}{Ar(g)}.$$

Several other types of redox reactions do not fall in the oxygen-atom, hydrogen-atom, or electron-transfer categories. Important among these are reactions of fluorine, chlorine, bromine, and iodine. These four elements, known as the halogens, form diatomic molecules, which are versatile oxidizing agents. The following examples are typical:

$$\underset{\text{fluorine}}{3F_2} + \underset{\text{sulfur}}{S} \longrightarrow \underset{\text{sulfur hexafluoride}}{SF_6}$$

$$\underset{\text{chlorine}}{Cl_2} + \underset{\text{iron(II) ion}}{2Fe^{2+}(aq)} \longrightarrow \underset{\text{chloride ion}}{2Cl^-} + \underset{\text{iron(III) ion}}{2Fe^{3+}(aq)}$$

$$\underset{\text{bromine}}{Br_2} + \underset{\text{ethylene}}{H_2C=CH_2} \longrightarrow \underset{\text{ethylene dibromide}}{BrCH_2CH_2Br}$$

$$\underset{\text{iodine}}{I_2} + \underset{\text{thiosulfate ion}}{2S_2O_3^{2-}} \longrightarrow \underset{\text{iodide ion}}{2I^-} + \underset{\text{polythionate ion}}{S_4O_6^{2-}}.$$

Such reactions often qualify as redox processes only in the broad sense that oxidation-state changes occur. The oxidation-state characterization extends oxidation–reduction chemistry to include examples from the reactions of all the chemical elements.

The Earth's surface as a redox boundary

Significance of redox reactions. Oxidation–reduction reactions have vast importance not only in chemistry but in geology and biology as well. The surface of the Earth is a redox boundary between the planet's reduced metallic core and an oxidizing atmosphere. The Earth's crust is largely composed of metal oxides, and the oceans are filled with water, an oxide of hydrogen. The tendency of nearly all surface materials to be oxidized by the atmosphere is reversed by the life process of photosynthesis. Because they are constantly renewed by the photosynthetic reduction of carbon dioxide, life's complex compounds can continue to exist on the Earth's surface.

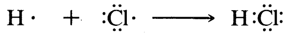
For similar reasons, much of chemical technology hinges on the reduction of materials to oxidation states lower than those that occur in nature. Such basic chemical products as ammonia, hydrogen, and nearly all the metals are produced by reductive industrial processes. When not used as structural materials, these products are reoxidized in their commercial applications. The weathering of materials, including wood, metals, and plastics, is oxidative, since, as the products of technological or photosynthetic reductions, they are in oxidation states lower than those stable in the atmosphere.

Solar radiation is converted to useful energy by a redox cycle that operates continually on a global scale. Photosynthesis converts radiant energy into chemical potential energy by reducing carbon compounds to low oxidation states, and this chemical energy is recovered either through enzymatic oxidations at ambient temperatures or during combustion at elevated temperatures.

THEORETICAL ASPECTS

Diagramming electron-pair bonding

Oxidation states. The idea of assigning an oxidation state to each of the atoms in a molecule evolved from the electron-pair concept of the chemical bond. Atoms within a molecule are held together by the force of attraction that the nuclei of two or more of them exert on electrons in the space between them. In many cases this sharing of electrons can be regarded as involving electron-pair bonds between adjacent nuclei. Electron-pair bonding is often diagrammed so as to show all the bonding and nonbonding valence electrons; *e.g.,* the structures of atomic hydrogen, atomic chlorine, and hydrogen chloride shown below (each dot represents one valence electron):



$$\underset{\text{hydrogen atom}}{H\cdot} + \underset{\text{chlorine atom}}{:\underset{\cdot\cdot}{\ddot{Cl}}\cdot} \longrightarrow \underset{\text{hydrogen chloride molecule}}{H:\underset{\cdot\cdot}{\ddot{Cl}}:}$$

The hydrogen chloride diagram reflects the presence, in the internuclear region, of two electrons that are under the mutual attractive influence of both the hydrogen and chlorine nuclei. Oxidation states for the hydrogen and chlorine in HCl are assigned according to the net charges that remain on H and Cl when the shared electrons are assigned to the atom that has the greater attraction for them. Through physical measurements on isolated atoms and simple molecules, these relative attractive powers have

Table 3: Pauling Electronegativities of Selected Elements

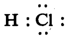
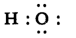
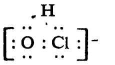
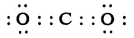
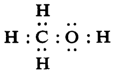
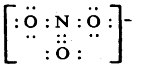
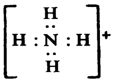
Fluorine	4.0
Oxygen	3.5
Nitrogen	3.0
Chlorine	3.0
Bromine	2.8
Sulfur	2.5
Iodine	2.5
Carbon	2.5
Hydrogen	2.1
Phosphorus	2.1
Iron	1.8
Sodium	0.9

Source: L. Pauling, *The Nature of the Chemical Bond.*

been determined. Table 3 lists the electronegativity values for some important elements.

In the hydrogen chloride molecule the chlorine is more electronegative than hydrogen and is, therefore, assigned both shared electrons. Chlorine has seven valence electrons in its neutral state. Having acquired an eighth electron in its reaction with hydrogen, chlorine is considered to have an oxidation state of −1. Hydrogen, on the other hand, is assigned +1, having lost the single valence electron that it has in its neutral state. Charges arrived at in this way are the basis for oxidation-state assignments, conventionally represented by roman numerals, such as in H(I) and Cl(−I) for the constituents of HCl. Because determination of oxidation states is simply a method of conceptually distributing shared electrons to individual atoms, the same number of electrons must be accounted for, before and after such assignment. Table 4 includes examples of molecules that have multiple bonds. The oxidation states of the atoms involved are added up algebraically in the table, and their sum must always equal the net charge on the molecule. There is no physical reality to oxidation states; they simply represent the results of calculations based on a formal rule.

Table 4: The Oxidation States of the Atoms in Typical Small Molecules

molecular species	oxidation-state assignments	algebraic sums
	H(I), Cl(−I)	1 − 1 = 0
	H(I), O(−II)	2(1) − 2 = 0
	Cl(I), O(−II)	1 − 2 = −1
	C(IV), O(−II)	4 − 2(2) = 0
	C(−II), O(−II), H(I)	4(1) − 2 − 2 = 0
	N(V), O(−II)	5 − 3(2) = −1
	H(I), N(−III)	4(1) − 3 = +1

Guidelines for assigning oxidation states

Oxidation states can be assigned for most common molecules with the help of a few guidelines. First, electrons shared by two atoms of the same element are divided equally; accordingly, elements are always in oxidation state of 0, regardless of their allotropic form (allotropic refers to the phenomenon of an element's having two or more forms; *e.g.,* carbon can exist as diamond or graphite and in both cases is in the 0 oxidation state). Second, only fluorine is more electronegative than oxygen. Therefore, except in compounds containing oxygen–oxygen or oxygen–fluorine bonds, oxygen can be reliably assigned the oxidation state −2. Similarly, hydrogen is less electronegative than fluorine, oxygen, nitrogen, chlorine, sulfur, and carbon (F, O, N, Cl, S, and C), so it is in the +1 oxidation state in its combinations with those elements. For many common compounds containing only hydrogen, oxygen, and a third element, the third element's oxidation state can be calculated, assuming oxidation numbers of +1 for hydrogen and −2 for oxygen. When bonds are present between two elements that differ little in electronegativity, however, oxidation-state assignments become doubtful, and the distinction between redox and nonredox processes is not evident.

There is a general reluctance, particularly regarding organic systems, to assume oxidation-state changes when the reaction results can be accounted for by the transfer or addition of water (H_2O), ammonia (NH_3), the hydroxide ion (OH^-), or the ions of hydrogen (H^+), chlorine (Cl^-), bromine (Br^-), or iodine (I^-), or combinations of these species; *e.g.,* the ammonium ion (NH_4^+), hydrogen chloride (HCl). The reason is that, in these molecules and ions, the elements are present in their most typical oxidation states: hydrogen(I), chlorine(−I), oxygen(−II), bromine (−I), iodine(−I), and nitrogen(−III).

Example of oxygen-atom-transfer reaction

The oxidation-state concept clarifies the relationship between oxygen-atom, hydrogen-atom, and electron transfer. The oxygen- and hydrogen-transfer criteria apply only when oxygen and hydrogen occur in their typical oxidation states. An example of an appropriate reaction involving oxygen-atom transfer is the reduction of ferrous oxide by carbon monoxide:

$$\begin{array}{ccccccc} \mathrm{FeO} & + & \mathrm{CO} & \longrightarrow & \mathrm{CO_2} & + & \mathrm{Fe.} \\ \text{iron(II) oxide} & & \text{carbon} & & \text{carbon} & & \text{iron} \\ \text{(ferrous oxide)} & & \text{monoxide} & & \text{dioxide} & & \end{array}$$

In terms of oxidation-state changes, this oxygen-atom transfer is equivalent to the two-electron reduction of iron and complementary two-electron oxidation of carbon:

$$\mathrm{Fe(II)} + \mathrm{C(II)} \longrightarrow \mathrm{C(IV)} + \mathrm{Fe(0)}.$$

Oxygen, which occurs in the oxidation state −2 in both reactants and products in the first equation, is not shown in the second. In transferring, the oxygen atom leaves two electrons behind, causing the reduction of iron, and acquires two electrons from the carbon atom, oxidizing the carbon.

In a similar way, the hydrogenation of ethylene corresponds to a two-electron reduction of the two-carbon skeleton:

$$\begin{array}{cccc} \mathrm{H_2} + \mathrm{H_2C{=}CH_2} & \longrightarrow & \mathrm{H_3CCH_3} \\ \text{hydrogen ethylene} & & \text{ethane} \end{array}$$

$$\begin{array}{ccc} \mathrm{2H(0)} & + & \mathrm{2C(-II)} \\ \text{hydrogen in} & & \text{carbon in} \\ \text{oxidation state zero} & & \text{oxidation state} - 2 \end{array}$$

$$\begin{array}{cccc} \longrightarrow & \mathrm{2H(I)} & + & \mathrm{2C(-III).} \\ & \text{hydrogen in} & & \text{carbon in} \\ & \text{oxidation state} + 1 & & \text{oxidation state} - 3 \end{array}$$

In this example also, the second equation includes only the atoms that change oxidation states: the four hydrogen atoms initially present in ethylene are in the +1 oxidation state in both reactants and products and are therefore omitted. Each of the two neutral hydrogen atoms can be regarded as giving up an electron to, and thereby reducing, one of the carbon atoms. This example also demonstrates clearly that the oxidation that complements the reduction of ethylene is that of the two hydrogen atoms in H_2—namely, from the 0 to the +1 oxidation state. General application of the oxidation-state concept leads to a formal viewpoint toward all redox reactions as electron-transfer reactions.

Half reactions. One of the basic reasons that the concept of oxidation–reduction reactions helps to correlate chemical knowledge is that a particular oxidation or reduction can often be carried out by a wide variety of oxidizing or reducing agents. Reduction of the iron(III) ion to the iron(II) ion by four different reducing agents provides an example:

$$\underset{\text{iron(III) ion}}{\mathrm{Fe^{3+}(aq)}} + \underset{\text{vanadium(II) ion}}{\mathrm{V^{2+}(aq)}} \longrightarrow \underset{\text{iron(II) ion}}{\mathrm{Fe^{2+}(aq)}} + \underset{\text{vanadium(III) ion}}{\mathrm{V^{3+}(aq)}}$$

$$\underset{\text{iron(III) ion}}{\mathrm{2Fe^{3+}(aq)}} + \underset{\text{zinc metal}}{\mathrm{Zn(s)}} \longrightarrow \underset{\text{iron(II) ion}}{\mathrm{2Fe^{2+}(aq)}} + \underset{\text{zinc(II) ion}}{\mathrm{Zn^{2+}(aq)}}$$

$$\underset{\text{iron(III) ion}}{\mathrm{2Fe^{3+}(aq)}} + \underset{\text{iodide ion}}{\mathrm{2I^-}} \longrightarrow \underset{\text{iron(II) ion}}{\mathrm{2Fe^{2+}(aq)}} + \underset{\text{iodine}}{\mathrm{I_2}}$$

$$\underset{\text{iron(III) ion}}{\mathrm{2Fe^{3+}(aq)}} + \underset{\text{hydroquinone}}{\mathrm{C_6H_4(OH)_2}} \longrightarrow \underset{\text{iron(II) ion}}{\mathrm{2Fe^{2+}(aq)}} + \underset{\text{quinone}}{\mathrm{C_6H_4O_2}} + \underset{\text{hydrogen ion}}{\mathrm{2H^+}}.$$

Production of the same change in the aqueous iron(III) ion by different reductants emphasizes the fact that the reduction is a characteristic reaction of the iron system itself, and, therefore, the process may be written without specifying the identity of the reducing agent in the following way:

$$\underset{\text{iron(III) ion}}{\mathrm{Fe^{3+}(aq)}} + \underset{\text{electron}}{e^-} \longrightarrow \underset{\text{iron(II) ion}}{\mathrm{Fe^{2+}(aq)}}.$$

Hypothetical equations of this type are known as half reactions. The symbol e^-, which stands for an electron, serves as a reminder that an unspecified reducing agent is required to bring about the change. Half reactions can be written, equally, for the reducing agents in the four reactions with ferric ion:

$$\underset{\text{vanadium(II) ion}}{\mathrm{V^{2+}(aq)}} \longrightarrow \underset{\text{vanadium(III) ion}}{\mathrm{V^{3+}(aq)}} + \underset{\text{electron}}{e^-}$$

$$\underset{\text{zinc(0) metal}}{\mathrm{Zn(s)}} \longrightarrow \underset{\text{zinc(II) ion}}{\mathrm{Zn^{2+}(aq)}} + \underset{\text{electron}}{2e^-}$$

$$\underset{\text{iodide ion}}{\mathrm{2I^-}} \longrightarrow \underset{\text{iodine}}{\mathrm{I_2}} + \underset{\text{electron}}{2e^-}$$

$$\underset{\text{hydroquinone}}{\mathrm{C_6H_4(OH)_2}} \longrightarrow \underset{\text{quinone}}{\mathrm{C_6H_4O_2}} + \underset{\text{electron}}{2e^-} + \underset{\text{hydrogen ion}}{\mathrm{2H^+}}$$

Although hypothetical, half reactions are properly balanced chemical processes. Since $\mathrm{V^{2+}(aq)}$ increases its oxidation number by one, from +2 to +3, in the first half reaction, an electron is shown as a product of the change. Similarly, two electrons are produced when the oxidation number of zinc increases from 0 to +2 in the second half reaction. When half reactions for hypothetical isolated oxidations and reductions are combined, the electrons must cancel if the equation for a possible overall chemical reaction is to result.

The use of half reactions is a natural outgrowth of the application of the electron-transfer concept to redox reactions. Since the oxidation-state principle allows any redox reaction to be analyzed in terms of electron transfer, it follows that all redox reactions can be broken down into a complementary pair of hypothetical half reactions. Electrochemical cells (in which chemical energy can be converted to electrical energy, and vice versa) provide some physical reality to the half-reaction idea. Oxidation and reduction half reactions can be carried out in separate compartments of electrochemical cells, with the electrons flowing through a connecting wire and the circuit completed by some arrangement for ion migration between the two compartments (but the migration need not involve any of the materials of the oxidation–reduction reactions themselves).

Redox potentials for common half reactions. The analysis of the electrical potential, or voltage, developed by pairing various half reactions in electrochemical cells has led to the determination of redox potentials for a substantial number of common half reactions. While a detailed description of redox potentials requires the methods of thermodynamics (the branch of physics concerned with the role played by heat in the transformation of matter or energy), a great deal of useful information can be obtained from redox potentials with minimal recourse to formal theory. Basically, a table of half-cell potentials is a summary of the relative tendencies of different oxidations and reductions to occur. Table 5 lists selected half reactions and their corresponding reduction potentials (which are symbolized by $E°$).

The physical significance of the values is directly linked to several agreements about their use. First, the greater the value of $E°$ (the reduction potential), the greater the tendency of a half reaction to proceed from left to right (as written). The half reactions in Table 5 are listed from top to bottom in order of decreasing $E°$: the higher a reaction's position on the list, the greater the tendency of the reactants to accept electrons. In other words, reagents high on the list, such as fluorine gas ($\mathrm{F_2}$) and permanganate ion ($\mathrm{MnO_4^-}$), are strong oxidizing agents. Second, the reduction of hydrogen ions ($\mathrm{H^+}$) to hydrogen gas ($\mathrm{H_2}$) is arbitrarily assigned the value 0 volts. Half cells with positive reduction potentials involve reactants that are more readily reduced than $\mathrm{H^+}$; conversely, those with negative potentials involve reactants that are more difficult to reduce than hydrogen ions.

Table 5: Selected Values of Standard Reduction Potentials

half reactions*	$E°$ (volts)
$\mathrm{F_2(g)} + 2e^- \rightarrow \mathrm{2F^-}$ fluorine(0) electrons fluoride(−I) ion	2.87
$\mathrm{MnO_4^-} + \mathrm{8H^+} + 5e^- \rightarrow \mathrm{Mn^{2+}(aq)} + \mathrm{4H_2O}$ permanganate ion hydrogen(I) ion electrons manganese(II) ion water	1.51
$\mathrm{Cl_2(g)} + 2e^- \rightarrow \mathrm{2Cl^-}$ chlorine(0) electrons chloride(−I) ions	1.36
$\mathrm{O_2(g)} + \mathrm{4H^+} + 4e^- \rightarrow \mathrm{2H_2O}$ oxygen(0) hydrogen(I) ions electrons water	1.23
$\mathrm{Fe^{3+}(aq)} + e^- \rightarrow \mathrm{Fe^{2+}(aq)}$ iron(III) ion electron iron(II) ion	0.77
$\mathrm{Cu^{2+}(aq)} + 2e^- \rightarrow \mathrm{Cu(s)}$ copper(II) ion electrons copper(0)	0.34
$\mathrm{2H^+} + 2e^- \rightarrow \mathrm{H_2(g)}$ hydrogen(I) ion electrons hydrogen(0)	0.00
$\mathrm{Zn^{2+}(aq)} + 2e^- \rightarrow \mathrm{Zn(s)}$ zinc(II) ion electrons zinc(0)	−0.76
$\mathrm{Na^+} + e^- \rightarrow \mathrm{Na(s)}$ sodium(I) ion electron sodium(0)	−2.71

*The identifications in parentheses refer to the physical state of the substance: (g), gas; (aq), as a hydrated positive ion in water; (s), as the pure solid.
Source: W. Latimer, *Oxidation Potentials.*

Use of reduction potentials to predict reactions

With the aid of reduction potentials, it is possible to predict whether a particular oxidation–reduction reaction can occur. The predictions require breaking down the overall reaction into two half reactions of known reduction potentials. For example, if a strip of zinc metal is dipped into a solution containing copper(II) ion, the possibility exists for a redox process, which can be regarded as the sum of the half reactions aqueous zinc ion ($\mathrm{Zn^{2+}}$[aq]) to zinc metal (Zn[s]) and aqueous copper ion ($\mathrm{Cu^{2+}}$[aq]) to copper metal (Cu[s]), as follows:

$$\underset{\text{zinc(0) metal}}{\mathrm{Zn(s)}} + \underset{\text{copper(II) ion}}{\mathrm{Cu^{2+}(aq)}} \longrightarrow \underset{\text{zinc(II) ion}}{\mathrm{Zn^{2+}(aq)}} + \underset{\text{copper(0) metal}}{\mathrm{Cu(s)}}.$$

Combining these two half reactions requires writing the zinc ion to zinc metal half reaction the reverse of the way it appears in Table 3. When the direction of a half reaction is reversed, so that it can be added to another half reaction, the sign of its redox potential is also reversed (in this case, from negative to positive), and the two reduction potential values are then added.

$$\begin{array}{rcll}
Cu^{2+}(aq) + 2e^- & \longrightarrow & Cu(s) & +0.34\ \text{volt} \\
Zn(s) & \longrightarrow & Zn^{2+}(aq) + 2e^- & +0.76 \\
\hline
Cu^{2+}(aq) + Zn(s) & \xrightarrow{\text{spontaneous}} & Cu(s) + Zn^{2+}(aq) & +1.10 \\
\begin{array}{cc}\text{copper(II)} & \text{zinc} \\ \text{ion} & \text{metal}\end{array} & & \begin{array}{cc}\text{copper} & \text{zinc(II)} \\ \text{metal} & \text{ion}\end{array} &
\end{array}$$

The resulting $E°$ value for the net reaction, +1.10 volts, measures the tendency of the net reaction to occur. If $E°$ for a particular net reaction is positive, the process may be expected to occur spontaneously when the reactants are mixed at specified concentrations (one mole per litre; see below *Oxidation–reduction equilibria*). Therefore, it is predicted that copper metal should be deposited on a strip of zinc metal when the latter is immersed in a solution of a copper(II) salt. This reaction is, in fact, readily observed in the laboratory. A more specific physical interpretation of the +1.10 volt value is that it represents the voltage that would be produced by an ideal electrochemical cell based on the copper(II) ion to copper metal and zinc(II) ion to zinc metal half reactions with all the reagents at specified concentrations.

When the same two half cells are combined, with both their directions (and therefore the signs of their redox potentials) reversed, it is predicted that the reverse reaction, the depositing of zinc metal from a zinc(II) ion solution onto a copper strip, will not occur spontaneously. As in the case of $E°$ values for half reactions, those for net redox reactions also change sign when the direction of the reaction is reversed.

$$\begin{array}{rcll}
Cu(s) & \longrightarrow & Cu^{2+}(aq) + 2e^- & -0.34\ \text{volt} \\
Zn^{2+}(aq) + 2e^- & \longrightarrow & Zn(s) & -0.76 \\
\hline
Zn^{2+}(aq) + Cu(s) & \xrightarrow{\text{nonspontaneous}} & Cu^{2+}(aq) + Zn(s) & -1.10
\end{array}$$

The results of the copper–zinc system can be applied more generally to the half reactions in Table 3. For example, copper(II) ion in water (Cu^{2+}[aq]) is an oxidant strong enough to force a half reaction lower on the table to proceed spontaneously in the opposite direction of that written. Therefore, not only is copper(II) ion expected to oxidize zinc metal (Zn[s]) to zinc(II) ion (Zn^{2+}[aq]); it is also predicted to oxidize hydrogen gas (H_2[g]) to hydrogen ion (H^+) and sodium metal (Na[s]) to sodium ion (Na^+). Similarly, fluorine gas (F_2[g]), the strongest oxidant listed in Table 3, is predicted to oxidize spontaneously the products of all the other half reactions in the table. In contrast, the strongest reducing agent is solid sodium metal (Na[s]), and it is expected spontaneously to reduce the reactants of all the other half cells.

Oxidation–reduction equilibria. In practice many chemical reactions can be carried out in either direction, depending on the conditions. The spontaneous direction predicted for a particular redox reaction by half-cell potentials is appropriate to a standard set of reaction conditions. Specifically, the temperature is assumed to be 25° C with reagents at specified concentrations. Gases are present at one atmosphere pressure and solutes at one mole per litre (one molecular weight in grams dissolved in one litre of solution) concentration (1M). Solids are assumed to be in contact with the reaction solution in their normal stable forms, and water is always taken to be present as the solvent. Many practical problems can be solved directly with standard reduction potentials.

Role of the Nernst equation

The usefulness of reduction potentials is greatly extended, however, by a thermodynamic relationship known as the Nernst equation, which makes it possible to calculate changes in half-cell potentials that will be produced by deviations from standard concentration conditions. In the reaction between zinc metal and copper(II) ion, standard conditions for zinc and copper metal require simply that both solids be present in contact with the solution; the $E°$ values are not affected by either the total or proportionate amounts of the two metals. The calculation that the overall reaction is spontaneous by +1.10 volts is based on standard one mole per litre (1M) concentrations for aqueous zinc(II) ion (Zn^{2+}[aq]) and aqueous copper(II) ion (Cu^{2+}[aq]). Using the Nernst equation it is found that $E°$ for the overall reaction will be +1.10 volts as long as both ions are present in equal concentrations, regardless of the concentration level.

On the other hand, if the ratio of the zinc(II) to copper(II) ion concentrations is increased, the reduction potential ($E°$) falls until, at a very high preponderance of zinc ion, $E°$ becomes 0 volt. At this point, there is no net tendency for the reaction to proceed spontaneously in either direction. If the zinc(II) to copper(II) ion ratio is increased further, the direction of spontaneity reverses, and zinc ion spontaneously oxidizes copper metal. In practice, such high zinc(II) to copper(II) ion concentration ratios are unattainable, which means that the reaction can only be carried out spontaneously with copper(II) ion oxidizing zinc metal. Many reactions with $E°$ values smaller than +1.10 volts under standard conditions can be carried out in either direction by adjusting the ratio of product and reactant concentrations. The point at which $E° = 0$ volt represents a state of chemical equilibrium. When chemical reactions are at equilibrium, the concentrations of the reagents do not change with time, since net reaction is not spontaneous in either direction. Measurements of half-cell potentials combined with Nernst-equation calculations are a powerful technique for determining the concentration conditions that correspond to chemical equilibrium.

Reaction rates. *Predictability.* There are practical limitations on predictions of the direction of spontaneity for a chemical reaction, the most important arising from the problem of reaction rates. An analogy can be made with the simple physical system of a block on a sloping plane. Because of the favourable energy change, the block tends spontaneously to slide down, rather than up, the slope, and, at mechanical equilibrium, it will be at the bottom of the slope, since that is the position of lowest gravitational energy. How rapidly the block slides down is a more complex question, since it depends on the amount and kind of friction present. The direction of spontaneity for a chemical reaction is analogous to the downhill direction for a sliding block, and chemical equilibrium is analogous to the position at the bottom of the slope; the rate at which equilibrium is approached depends on the efficiency of the available reaction processes. Between zinc metal and aqueous copper(II) ion, the reaction proceeds without observable delay, but various other spontaneous redox processes proceed at imperceptibly slow rates under ordinary conditions.

Biological processes. A particularly significant illustration of the role of mechanisms in determining the rates of redox reactions concerns respiration, the central energy-producing process of life. Foodstuffs that are oxidized by molecular oxygen during respiration are quite unreactive with oxygen before ingestion. Such high-energy foods as grains and sugar can resist the atmosphere indefinitely but are rapidly converted to carbon dioxide and water through combination with oxygen during respiratory metabolism. The situation is exemplified by the behaviour of glucose at ambient temperatures.

The significance of the different rate behaviour of high-energy foods inside and outside the cell has been dramatized by Albert Szent-Györgyi, a Hungarian biochemist resident in the United States, a pioneering researcher into the chemical mechanism of respiration:

> You remember the exciting story of the grave of the Egyptian emperor. At its opening the breakfast of the emperor was found unburned though it had been exposed to the action of oxygen during several thousand years at a temperature that was not very different from 37° C [98.6° F]. Had the king risen and consumed his breakfast, as he had anticipated doing, the food would have been oxidized in no time, that is to say the cells of the emperor would have made reactions take place that would not run spontaneously (from Albert V. Szent-Györgyi, *On Oxidation, Fermentation, Vitamins, Health and Disease;* the Williams and Wilkins Company, 1939).

Living systems are able to use respiratory oxidation as an energy source only because the same reactions are slow outside the cell. In return for providing an efficient mechanism for the oxidation of foods, the cell gains control over the disposition of the liberated chemical energy.

Examples such as the chemistry of respiration make clear the importance of determining the rates and mechanisms of redox reactions. Often questions are difficult to answer even in regard to relatively simple reactions. It has been pointed out that many redox processes can be categorized as oxygen-atom-, hydrogen-atom-, or electron-transfer processes. These categories describe the net changes that are involved but provide no insight into the mechanisms of the reactions.

Mechanisms of redox reactions. Some of the problems associated with formulating descriptions of the mechanisms are illustrated by the reaction between two metal ions that undergo complementary, one-unit changes in oxidation state:

$$M(\mathrm{III}) + N(\mathrm{II}) \longrightarrow M(\mathrm{II}) + N(\mathrm{III}).$$

There are many different metal ions, designated with the letters *M* and *N*, which participate in redox reactions with this basic stoichiometry. To define the possible mechanisms with any precision, it is necessary to identify the groups that are directly bonded to the metals, as well as to specify which particular metals take part in the reactions. One relatively simple example is the chromium(II)–iron(III) reaction. Both chromium and iron aquo ions are surrounded by six water molecules in both the +2 and +3 oxidation states.

$$\mathrm{Fe(H_2O)_6^{3+}} \quad + \quad \mathrm{Cr(H_2O)_6^{2+}}$$

hexaaquoiron(III) ion
hexaaquochromium(II) ion

$$\longrightarrow \quad \mathrm{Fe(H_2O)_6^{2+}} \quad + \quad \mathrm{Cr(H_2O)_6^{3+}}$$

hexaaquoiron(II) ion
hexaaquochromium(III) ion

The reaction appears from its stoichiometry to entail simple electron transfer from Cr(II) to Fe(III). There are, however, other possibilities. Their variety is most evident if it is assumed that the oxidizing agent, hexaaquoiron(III) ($\mathrm{Fe[H_2O]_6{}^{3+}}$), dissociates a hydrogen ion before reacting with hexaaquochromium(II) ($\mathrm{Cr[H_2O]_6{}^{2+}}$). The hexaaquoiron(III) ion is then converted to a hydroxopentaaquoiron(III) ion, known to be an important reactant in the Fe(III)–Cr(II) reaction. Its formation requires no change in the oxidation state of iron and should be regarded as a simple acid–base reaction. With hydroxopentaaquoiron(III) ion, $\mathrm{(H_2O)_5FeOH^{2+}}$, as the oxidizing agent, simple electron transfer does provide one plausible means for the reaction. All the steps in the mechanism must add up to give the overall reaction. Intermediate species such as hydroxopentaaquoiron(III) must cancel out in the addition process by occurring as products in one step and reactants in another:

$$\mathrm{Fe(H_2O)_6^{3+}} \longrightarrow \mathrm{(H_2O)_5FeOH^{2+}} + \mathrm{H^+}$$

hexaaquoiron(III) ion
hydroxopenta-aquoiron(III) ion
hydrogen ion

$$\mathrm{(H_2O)_5FeOH^{2+}} \quad + \quad \mathrm{Cr(H_2O)_6^{2+}}$$

hydroxopenta-aquoiron(III) ion
hexaaquo-chromium(III) ion

$$\xrightarrow[\text{transfer}]{\text{electron}} \mathrm{(H_2O)_5FeOH^{+}} \quad + \quad \mathrm{Cr(H_2O)_6^{3+}}$$

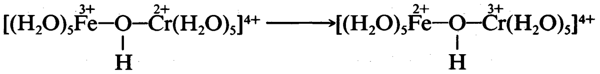
hydroxopenta-aquoiron(II) ion
hexaaquo-chromium(III) ion

$$\mathrm{(H_2O)_5FeOH^{+}} + \mathrm{H^+} \longrightarrow \mathrm{Fe(H_2O)_6^{2+}}$$

hydroxopenta-aquoiron(II) ion
hydrogen ion
hexaaquoiron(II) ion

$$\mathrm{Fe(H_2O)_6^{3+}} + \mathrm{Cr(H_2O)_6^{2+}} \longrightarrow \mathrm{Fe(H_2O)_6^{2+}} + \mathrm{Cr(H_2O)_6^{3+}}$$

An alternative to simple electron transfer is the transfer of a hydrogen atom from a water molecule bound to chromium(II) to the hydroxide ion ($\mathrm{OH^-}$) bound to iron(III). A third important possibility is the transfer of a neutral OH group (called a hydroxyl radical) from iron(III) to chromium(II). Oxygen and hydrogen occur as oxygen(−II) and hydrogen(I) in both reactants and products. These oxidation states correspond to those present in the hydroxide ion $\mathrm{OH^-}$. Therefore, when an OH group transfers, it must leave an electron behind, reducing the iron, and then reaccept an electron from chromium, oxidizing it. Since both the aqueous iron and chromium ions have a strong tendency to remain surrounded by six oxygen molecules, the most likely method of OH transfer would be the sharing of the transferred oxygen between the two metals at an intermediate stage of the reaction. But iron(III) is not a strong enough oxidizing agent to produce a significant concentration of neutral OH groups from $\mathrm{OH^-}$ ions. A more acceptable description of the OH transfer mechanism, therefore, is to regard $\mathrm{OH^-}$ as a bridging group through which electron transfer is accomplished as the two metal ion groups share a bonded oxygen atom.



$$\mathrm{[(H_2O)_5\overset{3+}{Fe}{-}\underset{\substack{|\\ H}}{O}{-}\overset{2+}{Cr}(H_2O)_5]^{4+}} \longrightarrow \mathrm{[(H_2O)_5\overset{2+}{Fe}{-}\underset{\substack{|\\ H}}{O}{-}\overset{3+}{Cr}(H_2O)_5]^{4+}}$$

Experimentally, it is difficult to distinguish between the alternative mechanisms posed above. Classification schemes for reaction mechanisms are of little value unless they offer alternatives that can be experimentally verified. For that reason the most successful system for classifying the mechanisms of redox reactions between metal ions has been the inner sphere–outer sphere dichotomy. Outer-sphere reactions are those that take place without breaking any bonds between a metal and a group such as water or hydroxide ion bound to it. Both the simple electron- and hydrogen-atom-transfer mechanisms are outer-sphere processes. If, on the other hand, the oxidation-state changes take place within an intermediate product in which the two metals are directly bonded to a common bridging group (such as shown in the equation above), the mechanism is inner-sphere. Formation of the intermediate requires loss of a bound water molecule by one of the metals. The inner sphere–outer sphere distinction can be tested experimentally in some redox systems. (M.V.O.)

COMBUSTION AND FLAME

Combustion is a chemical reaction between substances, usually including oxygen and usually accompanied by the generation of heat and light in the form of flame. The rate or speed at which the reactants combine is high, in part because of the nature of the chemical reaction itself and in part because more energy is generated than can escape into the surrounding medium, with the result that the temperature of the reactants is raised to accelerate the reaction even more. A familiar example is a lighted match. When a match is struck, friction heats the head to a temperature at which the chemicals react and generate more heat than can escape into the air, and they burn with a flame. If a wind blows away the heat or the chemicals are moist and friction does not raise the temperature sufficiently, the match goes out; properly ignited, the heat from the flame raises the temperature of a nearby layer of the matchstick and of oxygen in the air adjacent to it, and the wood and oxygen react in a combustion reaction. When equilibrium between the total heat energies of the reactants and the total heat energies of the products (including the actual heat and light emitted) is reached, combustion stops. Flames have a definable composition and a complex structure; they are said to be multiform and are capable of existing at quite low temperatures, as well as at extremely high temperatures. The emission of light in the flame results from the presence of excited particles and, usually, of charged atoms and molecules and of electrons.

General principles. Combustion encompasses a great variety of phenomena with wide application in industry, the sciences, professions, and the home, and the application is based on knowledge of physics, chemistry, and mechanics; their interrelationship becomes particularly evident in treating flame propagation.

In general terms, combustion is one of the most important of chemical reactions and may be considered a culminating step in the oxidation of certain kinds of substances.

Though oxidation was once considered to be simply the combination of oxygen with any compound or element, the meaning of the word has been expanded to include any reaction in which atoms lose electrons, thereby becoming oxidized. As has been pointed out, in any oxidation process the oxidizer takes electrons from the oxidizable substance, thereby itself becoming reduced (gaining electrons). Any substance at all can be an oxidizing agent. But these definitions, clear enough when applied to atomic structure to explain chemical reactions, are not as clearly applicable to combustion, which remains, generally speaking, a type of chemical reaction involving oxygen as the oxidizing agent but complicated by the fact that the process includes other kinds of reactions as well, and by the fact that it proceeds at an unusually fast pace. Furthermore, most flames have a section in their structure in which, instead of oxidations, reduction reactions occur. Nevertheless, the main event in combustion is the combining of combustible material with oxygen.

History. Combustion, fire, and flame have been observed and speculated about from earliest times. Every civilization had its own explanation for them. The Greeks interpreted combustion in terms of philosophical doctrines, one of which was that a certain "inflammable principle" was contained in all combustible bodies and this principle escaped when the body was burned to react with air. A generalization of the concept was provided by the phlogiston theory, formulated in the 17th century (see above). Treated at first as a purely metaphysical quality, phlogiston was conceived later as a material substance having weight, and, sometimes, negative weight. The inadequacy of the phlogiston theory became apparent only in the late 18th century when it proved unable to explain a host of new facts about combustion that were being observed for the first time as the result of increasing accuracy in laboratory experiments.

The English natural philosopher Sir Francis Bacon observed in 1620 that a candle flame has a structure at about the same time that Robert Fludd, an English mystic, described an experiment on combustion in a closed container in which he determined that an amount of air was used up thereby. A German, Otto von Guericke, using an air pump he had invented in 1650, demonstrated that a candle would not burn in a container from which the air had been pumped. Robert Hooke, an English scientist, in 1665 suggested that air had an active component that, upon heating, combined with combustible substances, giving rise to flame. Another idea ascribed the high temperature of flame to the fast motion of active air particles, and it was learned that sulfur mixed with nitre can burn in the absence of air (nitre is a compound of oxygen which releases oxygen to the sulfur).

The first approximation of the true nature of combustion was posited by Lavoisier, who discovered in 1772 that the products of burned sulfur or phosphorus, in effect their ashes, outweighed the initial substances, and postulated that the increased weight was due to their having combined with air. Interestingly, it was already known that metals transformed by heat to metallic ash weighed less than the metallic ash, but the theory was that in certain cases phlogiston in metals had a negative weight, and upon escaping during combustion, left the ash of the metal heavier than it had been with the phlogiston in it. Later, Lavoisier concluded that the "fixed" air that had combined with the sulfur was identical to a gas obtained by Priestley on heating the metallic ash of mercury—that is, the "ashes" obtained when mercury was burned could be made to release the gas with which the metal had combined. This gas was also identical to that described by the Swedish chemist, Carl Wilhelm Scheele, as an active fraction of air that sustained combustion. Lavoisier called the gas "oxygen."

Discovery of oxygen

Lavoisier's theory that combustion was a reaction between the burning substance and the gas oxygen, present only to a limited extent in the atmosphere, was based on scientific principles, the most important of which was the law of the conservation of matter (after Einstein's relativity theory, of matter and energy): the total amount of matter in the universe is constant. Even ancient philosophers had guessed this law and it was substantiated in the 17th century. Lavoisier also clarified the concept of element into a modern generalization, that it was a substance that could not be broken down, and this, too, supported his theory. Soon after, studies of gases by John Dalton, and the first table of atomic weights that Dalton compiled, as well as many new gases discovered by other scientists, were important in supporting not only Lavoisier's theory of combustion but his whole new system of chemistry based on accurate measurement. The discoveries of nitrogen and hydrogen in the latter half of the 18th century, added to the earlier discoveries of carbon dioxide and carbon monoxide, and the discovery that the composition of air is remarkably constant though it is a mixture, all supported Lavoisier's theory. The proper explanation of combustion, perhaps the oldest recognized chemical reaction, is usually said to have been a keystone in the development of modern science.

From 1815 to 1819, Sir Humphry Davy experimented on combustion, including measurements of flame temperatures, investigations of the effect on flames of rarefied gases, and dilution with various gases; he also discovered catalytic combustion—the oxidation of combustibles on a catalytic surface accompanied by the release of heat but without flame.

Despite these discoveries, the materialistic theory of combustion lacked a clear concept of energy and, therefore, of the critical role that energy considerations play in an accurate explanation of combustion. It was Sir Benjamin Thompson's experiments with heat in 1798 that revealed evidence for the concept of heat as a movement of particles. Development of a kinetic theory of gases, based on the premise that heat results from the motion of molecules and atoms, of thermodynamics, and of thermochemistry, all in the 19th century, finally elucidated the energy aspects of combustion.

Investigation of burning velocities, experiments on the order of events in the combustion of gas mixtures, and study of the breaking down of gas molecules by heat (thermal dissociation), in the last half of the 19th century, played a vital part in the refinement of theories concerning combustion mechanism. Studies of light emitted by flames led to its analysis in the spectroscope, a device that separates a mixture of light waves into the component waves, and to spectral analysis generally, including theories of atomic and molecular spectra, which, in turn, contributed to an understanding of the nature of flames. The Bunsen burner was also of importance in the study of flame structure. Progress in industry was a powerful stimulus in the search for clarification of flame phenomena. Explosion hazards in coal mines had drawn attention to flame propagation as far back as 1815, when Davy invented his safety lamp. In 1881 detonation was discovered and led at the beginning of the 20th century to a detonation theory, based on the assumption that a gas behaved as a fluid under certain conditions. Chemical kinetics after the 1930s became an indispensable part of flame propagation theory.

Physical and chemical aspects of combustion. *The chemical reactions.* Combustion, with rare exceptions, is a complex chemical process involving many steps that depend on the properties of the combustible substance. It is initiated by external factors such as heat, light, and sparks. The reaction sets in as the mixture of combustibles attains the ignition temperature, and several aspects of this step can be defined.

Ignition temperature

First, a relationship exists between the ignition temperature and the pressure of the mixture under specific conditions. Figure 7 shows the relationship for a mixture of hydrogen and oxygen. Only one temperature corresponds to a given pressure, whereas one or three pressures, called the explosion limits, may correspond to one temperature. The mechanism of the reaction determines the explosion limits: the reaction can proceed only when the steps in the sequence of reactions occur faster than the terminal steps. Thus, for combustion to be initiated with light, or with a spark, the light intensity or the spark energy must exceed certain minimal values.

The complexity of the combustion reaction mechanism and the rapidly varying temperatures and concentrations

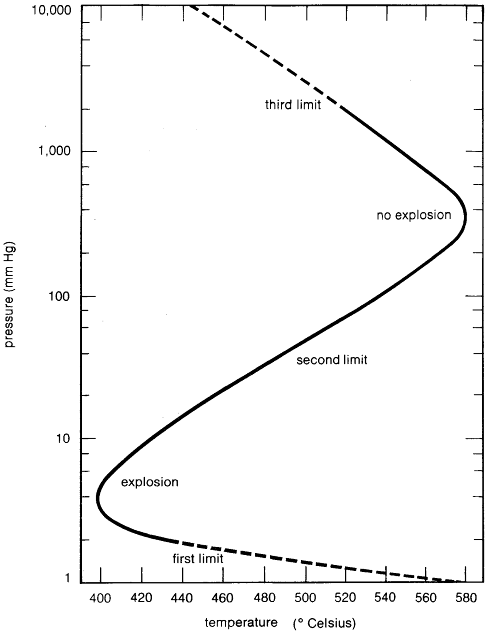


Figure 7: Explosion limits of a hydrogen-oxygen mixture.
From B. Lewis and G. von Elbe, *Combustion, Flames and Explosions of Gases* (1961); Academic Press, Inc.

in the mixture make it difficult and often impossible to derive an equation that would be useful for predicting combustion phenomena over wide temperature and concentration ranges. Instead, use is made of empirical expressions derived for specific reaction conditions.

Most reactions terminate when what is called thermal equilibrium has been attained—*i.e.,* when the energy of the reactants equals the energy of the products.

Special combustion reactions. Reactions of oxygen with hydrogen, with carbon monoxide, and with hydrocarbons are most important from the standpoint of theory and, at the same time, are of great practical value.

Hydrogen combustion proceeds by complicated branched-chain reactions involving the interaction of hydrogen and oxygen atoms with oxygen and hydrogen molecules, respectively, to produce hydroxyl radicals. The final reaction product is water, formed by the combination of hydroxyl with hydrogen molecules. Reactions that terminate the chain, such as recombination of atoms and fragments of molecules, and the adsorption of active particles on solid surfaces, also play an important part in the mechanism of hydrogen combustion. The knowledge of rate constants for these processes derived empirically makes possible a quantitative description of all combustion characteristics, such as explosion limits, delay of ignition, and burning velocity.

Combustion of carbon monoxide is mainly restricted to mixtures containing hydrogen or hydrogen compounds. In this case the reaction mechanism differs from that of hydrogen combustion chiefly in that it involves a step of fast interaction between hydroxyl and carbon monoxide.

Pure mixtures of carbon monoxide and oxygen (or air) can be ignited only with sparks of high energy, or under high pressures and temperatures. The chemical mechanism of their combustion is not yet clear, probably because oxidation of carbon monoxide, a reaction that is part of the combustion of practically all natural fuels, usually occurs in the presence of hydrogen or hydrogen compounds: the breakdown of wood, coal, petroleum, etc., during burning produces carbon monoxide, hydrogen, and compounds of carbon and hydrogen.

The mechanisms of combustion of hydrocarbons and of other organic compounds are known in general outline only. Many elementary steps of hydrocarbon combustion involving oxygen and hydrogen atoms, hydroxyl and organic radicals are similar to those for hydrogen, however, the overall mechanism of hydrocarbon combustion is complicated by the diversity of molecules and radicals involved. Moreover, with oxidation, thermal decomposition, a breakdown of complex organic compounds without oxidation, also takes place.

Cool and
hot flames

Two types of hydrocarbon combustion have been defined; (1) slow combustion at temperatures below 500° C, including cool flames observed at certain pressures, and (2) combustion at higher temperatures, accompanied by hot flames.

Ignition in two stages is characteristic of higher hydrocarbons in which a cool-flame stage yielding readily oxidizable products precedes that of the hot flame.

Physical processes. In addition to chemical reactions, physical processes that transfer mass and energy by diffusion or convection occur in gaseous combustion. In the absence of external forces, the rate of component diffusion depends upon the concentration of the constituents, pressure, and temperature changes, and on diffusion coefficients (a measure of the speed of diffusion). The latter are either measured or calculated in terms of the kinetic theory of gases. The process of diffusion is of great importance in combustion reactions, in flames—that is, in gaseous mixtures, and in solids or liquids. Diffusion heat transfer (by molecular means) follows a law (Fourier law) stating that the heat flux (a measure of the quantity of heat) is proportional to the temperature gradient (the difference in temperature between two limiting temperatures). The coefficient of proportionality, called the thermal conductivity coefficient, is also measured or calculated in terms of the kinetic theory of gases, like the diffusion coefficient.

Convection

Convective transport of mass and energy may be accounted for by buoyant forces, external forces, and turbulent and eddy motions. Convection is in the main responsible for the mixing of gases (*e.g.,* in furnaces).

Sublimation (the direct evaporation of a solid without intermediate liquid stage), evaporation of liquids, and the mechanical destruction of burning samples, all methods of transforming solids and liquids into gases, greatly contribute to their ease of combustion. In general, the combustion of condensed systems—*i.e.,* liquids and solids—is more complex than that of gases, which accounts for the greater development of the gas combustion theory.

Energy transport in combustion may also occur by the emission of light, mostly in the infrared.

Combustion phenomena and classification. All flames can be classified either as premixed flames or as flames that burn without premixing.

Premixed flames. Flame combustion is most prominent with fuels that have been premixed with an oxidant, either oxygen or a compound that provides oxygen, for the reaction. The temperature of flames with this mixture is often several thousand degrees. The chemical reaction in such flames occurs within a narrow zone several microns thick. This combustion zone is usually called the flame front.

Dilution of the burning mixture with an inert gas, such as helium or nitrogen, lowers the temperature and, consequently, the reaction rate. Great amounts of inert gas extinguish the flame, and the same result is achieved when substances that remove any of the active species are added to the flame. Conditions must be such that the flame is fixed at the burner nozzle or in the combustion chamber; this positioning is required in many practical uses of combustion. Various devices, such as pilot flames and recirculation methods, are designed for this purpose.

The principal quantitative characteristic of a flame is its normal, or fundamental, burning velocity, which depends on the chemical and thermodynamic properties of the mixture, and on pressure and temperature, under given conditions of heat loss. The burning velocity value ranges from several centimetres even to tens of metres, per second. The dependence of the burning velocity on molecular structure, which is responsible for fuel reactivity, is known for a great many fuel–air mixtures.

A widely applied thermal theory, one of the first flame propagation theories, implies that combustion proceeds primarily at temperatures close to the maximum the flame can achieve. A set of differential equations developed for thermal conductivity and diffusion is reduced to one equation that yields the burning velocity value. Further development of the theory has been made and computers now make most simplifications unnecessary.

According to thermal theories, flame propagation is accounted for by heat energy transport from the combustion zone to the unburned mixture, to raise the temperature of the mixture. Diffusion theory assumes that thermodynamic equilibrium sets in behind the flame front at a maximum temperature, and that radicals produced in this zone diffuse into the unburned mixture and ignite it. Both heat transport and diffusion of active particles must be considered essential for ignition.

Diffusion flames. Diffusion flames, smoothly flowing (laminar) or turbulent, belong to the class of flames whose ingredients are not mixed prior to entering the burning zone. Molecular or turbulent diffusion is responsible for the mixing of the gases in such flames. The distribution of the combustible material and of oxygen over various flame cross sections is regular in laminar flames but becomes much more complicated in turbulent gas flows. The transition from laminar to turbulent flames occurs at a certain point of the flow and depends upon the actual conditions.

Industrial flames

Industrial flames, including those activated in furnaces, belong to the turbulent diffusion type. The theory of diffusion flames, however, is less advanced than that of premixed flames and, since gas intermixing is mainly responsible for the structure and the features of diffusion flames, the theory operates more in terms of physics and of thermodynamics than chemical reactions. Flames such as those of candles, of liquid droplets, and of many propellants and condensed fuels are diffusion flames.

Diffusion flame

Oxidizing and reducing flame. When a premixed flame burns in open air with an excess of fuel, there appears in addition to the flame zone a zone of diffusion flame; this is accounted for by the diffusion of atmospheric oxygen, as, for example, in the Bunsen flame produced by a burner to which the air intake can be regulated, thereby altering the flow from an intensely hot one—in which most of the fuel gases are oxidized to carbon dioxide and water—to a relatively low temperature flux, in which most of the fuel gas is only partially oxidized. Such flames consist of an inner and outer cone—two zones in which different chemical reactions take place, the reducing and oxidizing zones, respectively. The oxidizing nature of the outer cone is due to excess oxygen (see Figure 8).

From A. Gaydon, *The Spectroscopy of Flames;* Chapman & Hall Ltd.

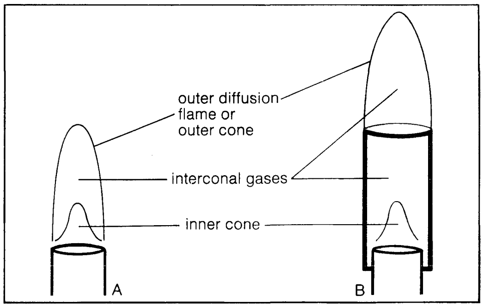


Figure 8: *Cones of burner flames.*
(A) Bunsen burner. (B) Teklu-Smithells burner.

Explosions. The transition from combustion to explosion is caused by an acceleration of the reaction, induced either by a rise in temperature or by increasing lengths of the reaction chain. The first is called thermal explosion, and the second, chain explosion.

Thermal explosions. Thermal explosion theory is based on the idea that progressive heating raises the rate at which heat is released by the reaction until it exceeds the rate of heat loss from the area. At a given composition of the mixture and a given pressure, explosion will occur at a specific ignition temperature that can be determined from the calculations of heat loss and heat gain.

The thermal explosion theory accounts for temperature rise and fuel consumption during the induction period. At sufficiently high rates of consumption the explosion will not occur.

Chain-branch reactions. It follows from the theory of branched-chain reactions that there is a limit to ignition, or to explosion, without a rise of temperature. In this case, what is called a chain explosion will occur when the probabilities of chain branching and of termination are equal. Usually, however, explosions are of a chain-thermal nature (*i.e.,* both heat accumulation and chain autoacceleration contribute to explosion).

Detonation. The progressive acceleration of reaction accounted for by growth of the flame front area and by transition from laminar to turbulent flow gives rise to a shock wave. The increase in temperature due to compression in the shock wave results in self-ignition of the mixture and detonation sets in. The shock wave-combustion zone complex forms the detonation wave. Detonation differs from normal combustion in its ignition mechanism and in the supersonic velocity of 2–5 kilometres per second for gases and 8–9 kilometres per second for solid and liquid explosives.

Detonation is impossible when the energy loss from the reaction zone exceeds a certain limit. The detonation limits observed for high explosives are also eventually dependent on factors responsible for the chemical reaction rate: the charge and diameter of the grain, etc.

Special aspects. Emission of light is a characteristic feature of combustion. Infrared, visible, and ultraviolet bands of molecules, and atomic lines are usually observed in flame spectra. Moreover, continuous spectra from incandescent particles, or from recombination of atoms, radicals, and ions are frequently observed. The sources of flame radiation are the thermal energy of gas (thermoluminescence) and the chemical energy released in exothermic elementary reactions (chemiluminescence). In a Bunsen burner fed with a sufficient amount of air, up to 20 percent of the reaction heat is released as infrared energy and less than one percent as visible and ultraviolet radiation, the infrared being mostly thermoluminescence. Radiation from the inner cone of the Bunsen flame in the visible and ultraviolet regions represents chemiluminescence.

Ionization

Many flames contain electrons and positive and negative ions, a fact evident from the electrical conductivity of flames, and also from the deviation of the flame cone in an electric field (the charges are attracted or repelled, distorting the flame), a phenomenon usually interpreted as a mechanical effect called electric wind. The resulting change of the flame shape can affect the burning velocity. Ionization, like the emission of light, can be the result of equilibrium processes, when it is called thermal ionization, or it can be related to chemical processes and called chemical ionization. Thermal ionization may be expected in very hot flames containing alkali metals or alkaline-earth metals (for example, sodium and calcium) as impurities because of their low-ionization potentials. The high concentrations of ions and electrons in the flames of organic species are undoubtedly due to chemical, rather than to thermal ionization. This is seen in the fact that electroconductivity in the inner cone of the Bunsen flame is several times higher than that of the outer cone. The reactions of ions and electrons may yield atoms and radicals and in this way affect the burning velocity.

Formation of soot is a feature of all hydrocarbon flames. It makes the flame luminous and nontransparent. The mechanism of soot formation is accounted for by simultaneous polymerization, a process whereby molecules or molecular fragments are combined into extremely large groupings, and dehydrogenation, a process that eliminates hydrogen from molecules.

Applications. The uses of combustion and flame phenomena can be categorized under five general heads.

In heating devices. Heating devices for vapour production (steam, etc.), in metallurgy, and in industry generally, utilize the combustion of gases, wood, coal, and liquid fuels. Control of the combustion process to obtain optimal efficiency is ensured by proper ratio and distribution of the fuel and the oxidant in the furnace, stove, kiln, etc., choice of conditions for heat transport from the combustion products to the heated bodies, and by appropriate aerodynamics of gas flows in the furnace. Radiation contributes to a certain extent to heat exchange. Combustion in furnaces being a very complicated process, only general ideas can be given by the combustion theory, so that the optimal conditions and the furnace design are usually decided empirically.

In explosives. The combustion and detonation of explosives are widely used in all sorts of work with mechanical action or explosion as the eventual goals. Practical applications of explosives are based on the theory of their combustion and detonation. The combustion of condensed explosives occurs mostly in the gas phase due to their evaporation, sublimation, or decomposition and can be treated in terms of the theory of gas combustion, which provides for the burning velocity, its dependence on temperature, pressure, and on the parameters determining the combustion regime and the nature of explosives, etc. Control of combustion and detonation in their practical applications is made possible by use of the theory, together with the results of experimental investigations on combustion and detonation.

In internal combustion engines. These comprise various engines, gas turbines, turbojets, and ramjets. The Otto engine operates with a mixture compressed in a cylinder by a piston. Shortly before the piston reaches the top the mixture is ignited with a spark, and the flame propagates at a normal velocity into the unburned mixture, increasing the pressure and moving the piston. There is a maximum of compression for any mixture composition and any engine design. Detonation occurs beyond this maximum because of the appearance of centres where self-ignition takes place before the flame front. Loss of power is one result of detonation; compounds hindering self-ignition are used to prevent it.

The diesel engine operates with a fuel spray injected into the engine cylinder as liquid droplets that mix with air by turbulent diffusion and that evaporate. At normal operations of the engine the temperature of compressed air is sufficiently high for self-ignition of the fuel.

In gas turbines, compressed air enters the combustion chamber where it mixes with the fuel. The expanding combustion products impart their energy to the turbine blades.

Two kinds of jet engines are used in aircraft: the turbojet and the ramjet. The turbine of a turbojet engine is used to operate the compressor. Thrust comes from the repulsion of products flowing out of a nozzle. In a ramjet engine, air is compressed and slowed down in the diffuser without any compressor or turbine device.

In rocket propulsion. The products of combustion of gaseous, liquid, or solid propellants in rockets are ejected from the combustion chamber through the (de Laval) nozzle at a high velocity. Knowledge of the kinetics of chemical processes in the nozzle is essential to determine the thrust required. The thrust decreases with the increasing mean molecular weight of the combustion products. Mixtures of low molecular weight and high heat of combustion, therefore, are used for rockets.

In chemical reactions. Flames are used in various ways to produce chemical reactions. The bead test in analytical chemistry is one example. The reducing power of a flame that has insufficient oxygen is utilized in limited ways. The soot produced by some flames is commercially useful. The manufacture of coke and charcoal is dependent on the judicious control of combustion and flame. (V.N.K.)

Electrochemical reactions

In the most general terms, electrochemical reactions are those chemical changes that are associated with the passage of electrons across interfaces between metal and solution. Under ordinary conditions, the occurrence of a chemical reaction is accompanied by the liberation or absorption of heat and not of any other form of energy; but there are many chemical reactions that—when allowed to proceed in contact with two electronic conductors, separated by conducting wires—liberate what is called electrical energy, and an electric current is generated. Conversely, the energy of an electric current can be used to bring about many chemical reactions that do not occur spontaneously. A process involving the direct conversion of chemical energy when suitably organized constitutes an electrical cell. A process whereby electrical energy is converted directly into chemical energy is one of electrolysis; *i.e.,* an electrolytic process. By virtue of their combined chemical energy, the products of an electrolytic process have a tendency to react spontaneously with one another, reproducing the substances that were reactants and were therefore consumed during the electrolysis. If this reverse reaction is allowed to occur under proper conditions, a large proportion of the electrical energy used in the electrolysis may be regenerated. This possibility is made use of in accumulators or storage cells, sets of which are known as storage batteries. The charging of an accumulator is a process of electrolysis; a chemical change is produced by the electric current passing through it. In the discharge of the cell, the reverse chemical change occurs, the accumulator acting as a cell that produces an electric current.

Finally, the passage of electricity through gases generally causes chemical changes, and this kind of reaction forms a separate branch of electrochemistry that will not be treated here.

GENERAL PRINCIPLES

Electronic and electrolytic conductors

Substances that are reasonably good conductors of electricity may be divided into two groups: the metallic, or electronic, conductors and the electrolytic conductors. The metals and many nonmetallic substances such as graphite, manganese dioxide, and lead sulfide exhibit metallic conductivity; the passage of an electric current through them produces heating and magnetic effects but no chemical changes. Electrolytic conductors, or electrolytes, comprise most acids, bases, and salts, either in the molten condition or in solution in water or other solvents. Plates or rods composed of a suitable metallic conductor dipping into the fluid electrolyte are employed to conduct the current into and out of the liquid; *i.e.,* to act as electrodes. When a current is passed between electrodes through an electrolyte, not only are heating and magnetic effects produced but also definite chemical changes occur. At or in the neighbourhood of the negative electrode, called the cathode, the chemical change may be either the deposition of a metal or the liberation of hydrogen and formation of a basic substance, or some other chemical reduction process; at the positive electrode, or anode, it may be either the dissolution of the anode itself, the liberation of a nonmetal, the production of oxygen and an acidic substance, or some other chemical oxidation process.

An electrolyte, prepared either by the melting of a suitable substance or by the dissolving of it in water or other liquid, owes its characteristic properties to the presence in it of electrically charged atoms or groups of atoms produced by the spontaneous splitting up or dissociation of the molecules of the substance. In solutions of the so-called strong electrolytes, most of the original substance, or in some solutions perhaps all of it, has undergone this process of electrolytic dissociation into charged particles, or ions. When an electrical potential difference (*i.e.,* a difference in degree of electrification) is established between electrodes dipping into an electrolyte, positively charged ions move toward the cathode and ions bearing negative charges move toward the anode. The electric current is carried through the electrolyte by this migration of the ions. When an ion reaches the electrode of opposite polarity, its electrical charge is donated to the metal, or an electric charge is received from the metal. The ion is thereby converted into an ordinary neutral atom or group of atoms. It is this discharge of ions that gives rise to one of the types of chemical changes occurring at electrodes.

HISTORY

The study of electrochemistry began in the 18th century, bloomed until the early 20th century, and then faded, owing to an excessive use of thermodynamic principles in analyzing the processes that take place at points in the system where the various parts form interfaces. Since about 1950 electrochemistry has undergone a change. The study of processes in solutions has been less stressed, but the study of the transfer of electrons between metals and solution has increased explosively. With this new emphasis electrochemistry is becoming a core science. It promises to be an important part of the foundation of the ecology-oriented society of the future, because electricity is not a pollutant. The pollution associated with some methods of generating electricity must, however, be reduced.

The first electrochemical reactions studied, in 1796, were those in the cell of silver and zinc plates with blotting paper wetted by aqueous salt solution between them; these cells were constructed by the Italian scientist Alessandro Volta, for whom the term volt was named. This cell was the first primary battery used for the production of electricity.

Faraday's laws

Michael Faraday formulated the laws of electrochemical stoichiometry, which deals with the application of laws of definite proportions and of the conservation of matter and energy to chemical activity. These state that a coulomb of electricity, a unit of charge, reacts with fixed quantities of a substance (*e.g.,* with 1.11800 milligrams of silver ions) or else that 1 gram equivalent of any substance reacts with 96,489 ± 2 coulombs. This latter number represents a fundamental quantity known as one faraday of electricity. The relationship between the chemical affinity of the reactants in the cell and the voltage of the cell when it is operating was precisely defined by the U.S. chemist Josiah Willard Gibbs in 1875 while the relation of this affinity to the potential of the electrochemical cell was initially formulated by the German physical chemist Walther Hermann Nernst in 1889.

The rise of electrodics

The period 1910 to 1950 was one of decline in electrochemistry, until it became limited mainly to the study of solutions. There was almost no progress in the understanding of electrochemical reactions outside of equilibrium conditions and reversibility, and knowledge of these was applied invalidly to reactions occurring at a net rate; *i.e.,* reactions not in equilibrium and not totally reversible. From about 1950 the study of electrified interfaces, with special reference to the study of the transfer of electrons (called electrodics), gained in importance and became the main aspect of electrochemistry. From about 1960, electrodics began to develop as an interdisciplinary area in the search for solutions to problems such as the source of energy in space flights from fuel cells, the stability of metals in moist environments, the electrochemical aspects of biological functions, extractions from mixtures, and the replacement of fossil fuels such as coal and petroleum and their by-products, by electricity produced or stored electrochemically in transportation.

THE ELECTROCHEMICAL PROCESS

Interactions of matter associated with the passage of an electric current depend upon the characteristics of the negatively charged electron. As the basic particle of electricity, the electron has an affinity for positively charged particles of matter, protons, whether in atoms, groups of atoms, or molecules. This affinity is analogous to the chemical affinity that particles exhibit among themselves. In fact, all chemical reactions result from a shift in the electron structure of atoms, and free electrons can combine with particles of matter (reduction) or be released by them (oxidation). The quantitative relationship between the free electrons of an electric current and the particles of a substance in which they cause a reaction is defined by the laws of Faraday (see above *History*). The substances that take part in electrochemical reactions called electrolytes, or ionic conductors, have been described above.

Electrons are available in large quantities in a relatively free (mobile) state only in substances called electronic conductors, among which metals are the most important. Thus, an electron conductor must be present as a basic component of any system in which electrochemical reactions are to occur. Furthermore, the availability of electrons in a conductor is limited by energy distribution to such an extent that electrochemical reactions take place only in the immediate vicinity of the electronic conductor's surface; *i.e.,* a few angstroms from the conductor into the solution. These reactions are, therefore, normally considered as occurring at the interface, or common boundary, between an electronic conductor, such as an electrode, and an ionic conductor of electricity, such as an electrolytic solution. Electrochemical reaction will take place, however, only to the extent that electricity can flow through such a system as a whole. To achieve this, it is necessary for the system to form a closed loop, electronically speaking. For schematic representation of these details, see Figure 9.

The closed loop of conductors

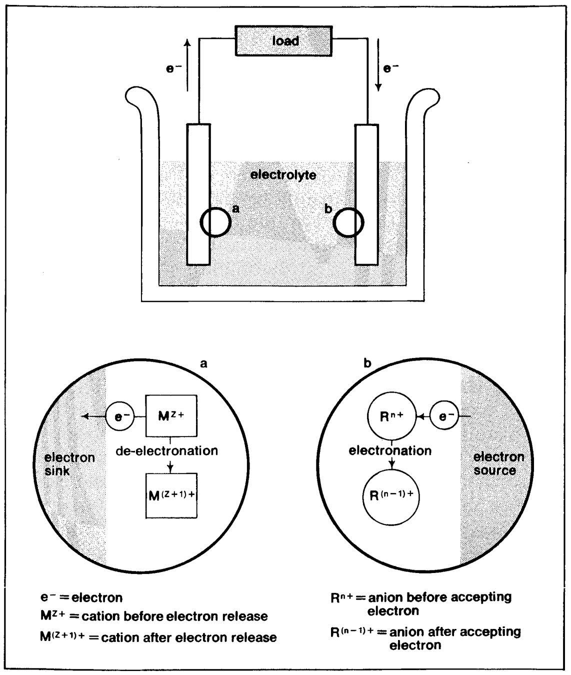


Figure 9: The mechanism of electron movement in an electrochemical cell.

To summarize, if at one metal-solution interface electrons are coming out of the metal, reducing a component of the solution, there must exist a second metal-solution interface where electrons are going into the metal in the process of oxidation.

The two electrodes and the ionic conductor in between (*e.g.,* an aqueous solution of some ionized salt) represent an electrochemical cell. The process occurring in the cell as a whole is a redox process with the reduction of one species spatially separated from the oxidation of another one. As a consequence of Faraday's law, the rates of electrochemical reactions at electrodes (expressed in terms of gram moles per second per square centimetre of electrode surface) are directly proportional to the current density (expressed in amperes per square centimetre); *i.e.,* current flowing through the cell divided by the electrode surface area.

Sites of electrochemical reactions. Electrochemical reactions take place where the electron conductor meets the ionic conductor; *i.e.,* at the electrode–electrolyte interface. Characteristic of this region, considered to be a surface phase, is the existence of a specific structure of particles and the presence of an electric field of considerable intensity (up to 10,000,000 volts per centimetre) across it; the field is caused by the separation of charges that are present between the two bulk phases in contact. For most purposes the surface phase can be considered as a parallel plate condenser, with one plate on the centre of the ions that have been brought to the electrode, at the distance of their closest approach to it, and with the second plate at the metal surface; between the two plates and acting as a dielectric (*i.e.,* a nonconducting material) are oriented water molecules. This structure is termed the electric double layer and is illustrated in Figure 10.

Thermal motion of the positive ions in the solution makes the condenser plate on the electrolyte side of the interface diffuse; *i.e.,* the ions are distributed in a cloudlike way. This condition justifies the division of the potential change between the bulk of metal and the bulk of electrolyte into two parts: first, that between the metal surface and the first ionic layer at the distance of closest approach (called the outer Helmholtz plane, in which the ions are usually surrounded by solvent particles; *i.e.,* are solvated); and second, that between the first ionic layer and the bulk of the solution, the diffuse part of the double layer. The picture is further complicated by the presence of ions in the electrode surface layer in addition to those that are present for electrostatic reasons; *i.e.,* by the force of attrac-

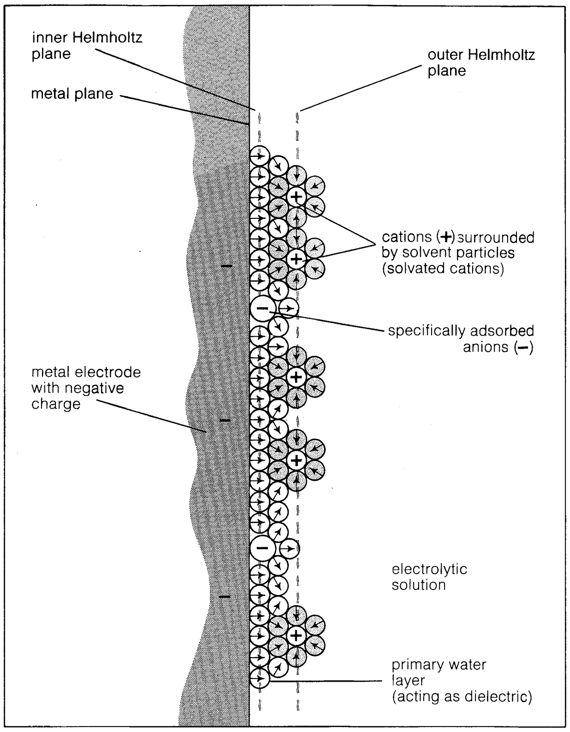


Figure 10: Double-layer structure and change of potential with distance from the electrode surface (see text).

tion or repulsion between electric charges. Such electrode surface layer ions are said to be specifically adsorbed on the electrode surface. Since this species of ions is attracted by the surface to a distance closer than the "distance of the closest approach" of ions, further subdivision of the inner part of the electric double layer is justified. Hence, the inner Helmholtz plane is introduced as the plane formed by the centres of specifically adsorbed ions. Adsorption of neutral molecules on the surface can also change the properties of the electric double layer. This change occurs as a consequence of replacing the water molecules, and thus changes that part of the potential (electrical) difference across the double layer that is caused by the adsorbed dipoles (water molecules that have a polarity—*i.e.,* they behave like minute magnets—because of their hydrogen-oxygen structure, making one end of the molecule positive and the other end negative).

Relative electrode potential

The absolute value of electrical potential difference, symbolized in calculation by the Greek letters delta and psi, $\Delta\psi$, between the bulk of a metal electrode and the bulk of an electrolyte cannot be measured. Instead, the voltage of a special cell, composed of the specific electrode being studied and of an arbitrarily selected reference electrode, is normally measured; the voltage is referred to as the relative electrode potential, E. Of special interest is that state of the electrode at which there is no net charge (in this case, no unbalanced, or extra positive, charge) at the metal side of the double layer. The relative potential at which this state is achieved is characteristic of each metal. This point is termed the potential of zero charge. At that potential, the field across the double layer is due to orientation of water molecules and other dipoles at the surface only.

Most of the knowledge of the detailed structure of the interface between a metal and an electrolyte arises from experimentation with mercury, the only metal that is liquid at ordinary temperatures; the double layer structure turns out to have surface tensions that must be measured, and this measurement is difficult with solid metals. By 1970, however, it had been shown that it is possible to measure surface tension changes at the metal-solution interface. Thus, the way to the determination of the double layer structure involving solids was opened.

Substances that are semiconductors can also be employed as electron carriers in electrochemical reactions. Semiconductors are substances which range between serving as insulators at low temperatures and as metallic-type conductors at high temperatures. In the case of semiconductors, however, the electric double layer has a more complex structure inasmuch as the condenser plate at the electrode side of the double layer also becomes diffuse. Thus, the overall potential difference between bulks of the phases in contact comprises also the potential difference between the bulk of the semiconductor and its surface.

Types of electrochemical reactions. There are several types of electrochemical reactions.

Simple redox reactions. A simple redox reaction is one that involves a change in the electrical charge of a charge carrier, usually a simple or complex ion in the solution, by its taking away, an electron from the electrode (reduction), or its giving an electron to the electrode (oxidation). The same carrier may be present in solution in two states of charge. The higher, more positive charge is called the oxidized state, and the lower, less positive charge is called the reduced state. For example, when ferric and ferrous ions are both present in solution in significant quantity, and when electron exchange with the electrode is sufficiently fast, redox equilibrium is established at the electrode, giving it a well-defined potential, or reversible redox potential.

Reversible hydrogen electrode

Reactions that produce gases. When hydrogen ions in solution react with electrons ejected from a metal, hydrogen atoms are formed at the surface, where they combine among themselves or with other hydrogen ions and electrons to give gaseous hydrogen molecules. If all the reactions are fast enough, an equilibrium is attained between hydrogen ions and gaseous hydrogen. A metal in contact with solution at which such a situation exists is called the reversible hydrogen electrode, and its electrical potential is arbitrarily taken to be zero; every other electrode can thus be compared with it as it represents the basis for constituting the hydrogen scale of relative electrode potentials. Similarly, negative hydroxyl ions in solution (OH^-) can be made to give up electrons to a metal and, in a series of reactions, the final one is the formation of gaseous oxygen. Chlorine is another gaseous product; it evolves upon electrochemical oxidation of chloride ions in concentrated solutions of neutral and acid salts.

Reactions that deposit and dissolve metals. When a metal ion is reduced and discharged as a neutral atom, or species, it tends to build into the metal lattice of the electrode. Thus, metals can be deposited at electrodes. Conversely, if electrons are taken away from the metal electrode by applying positive potentials to it, the metal ions thus formed can cross the double layer of electric charge at the interface, undergo hydration (combination with water), and enter the solution. The metal electrode thus dissolves. Many metals establish well-defined electric potentials when they are in contact with their own ions in solution.

Oxidation and reduction of organic compounds. A reaction of the oxidation and reduction of organic compounds can also be done at electrodes. Such reactions, however, are mostly irreversible in the literal sense that they lead to products that cannot easily be converted back into the original substance. Exceptions are some oxygen- and nitrogen-containing compounds (quinones, amines, and nitrous-compounds), that can give fairly well-defined reversible potentials.

Mechanism of charge transfer. The causes of the thermodynamically irreversible behaviour of electrode reactions are found in the nature of the elementary act of charge transfer. Like any chemical reaction, this act is inhibited by the existence of an energy barrier between the oxidized and the reduced state. This barrier implies that the reaction could take place only in the special circumstances when, during the course of numerous interactions with other species (atoms, ions, molecules, etc.) surrounding it, a molecule attains an excited state in which it has an abnormal energy content. In most chemical reactions, this energy content must be sufficient for the species to come into what is called the transition state; the transition state characterizes the top of the energy barrier just before a reaction begins. If such a model is applied to electron transfer at an interface, calculation shows that electron exchange reactions at electrodes would be prohibitively slow, a conclusion at variance with the observed phenomena; quantum mechanical laws, however, govern the motion of electrons, and their inclusion changes the calculations to fit reality. Quantum mechanics require that for fast electron exchange to take place, electrons in a particle outside the double layer (*e.g.,* a hydrated ion at the outer Helmholtz plane) must attain certain well-defined quantized energy levels equal to those in which free electrons exist in the metal. Since such states can be attained by the particle at a lower energy-content than that needed for its transfer over the top of the energy barrier, according to the classical view, this fast process of electron exchange between the electrode and a particle in solution is termed electron tunnelling through the energy barrier.

Electron tunnelling

Whereas the rate of chemical processes, or what may loosely be termed the speed of reaction, can be influenced only by changing the concentrations of reactants or by changing the temperature or both, the rate of electrochemical processes also can be manipulated by changing the electrode potential. Making the electrode more negative increases the number of electrons in the metal ready to tunnel to ions, and hence the rate of the reduction process increases. Conversely, making the potential more positive decreases this rate and increases the number of particles ready to give away electrons, thus increasing the rate of the oxidation process.

It can be deduced that there must exist a direct proportionality between the rate of reaction and the concentration of the reacting species and at the same time an exponential proportionality between the rate of reaction and the electrode potential.

At any electrode potential, both reduction of one species and oxidation of the product of reduction are taking place but at different rates; the rate of each reaction is determined by the respective concentration and by the corresponding effects of potential. The rate of an electrochemical reaction can best be described as the electric current density; *i.e.,* a measure of the quantity of electrons moving in a certain volume of space during a specified unit of time.

The relationships can be represented quantitatively by an equation in which the net, or resulting, current (the difference of the rate of electron ejection across the interface to particles in solution, diminished by the rate at which particles in solution inject electrons into the metal) is equated to the difference of the rates of reduction and oxidation and the variables and constants that relate to these reactions. See equation (1), and other equations below in *Calculations* at the end of this section. Equation (2), known as the Nernst equation, which can be derived from equation (1), gives the value of the electrode potential when the rate of oxidation exactly equals the rate of reduction. Using this value for the electrode potential, equation (3), called the Butler-Volmer equation, can be derived; it represents the most fundamental relationship in electrodic chemistry.

COMPLEX ELECTROCHEMICAL REACTIONS

Electrochemical processes considered so far involve simple reactions of a particle with a single electron to produce a reduced ion (*e.g.,* the ferrous ion of iron with two positive charges, Fe^{++}), or vice versa. Such are the simple ionic redox processes, where the only difference in structure between a reactant and a reaction product may be due to some rearrangement of the neighbouring solvent molecules. When one or more transfers of electrons between the electrode and a species in solution are accompanied by major structural changes (*e.g.,* when hydroxyl ions, OH^-, transform into a molecule of oxygen, O_2, and a water molecule, H_2O, in the process of oxygen evolution at the anode, or positive electrode), the reaction usually consists of a sequence of events, called elementary acts, or unit steps, constituting the reaction mechanism. Intermediate states between the steps usually involve some unstable intermediate species with higher energy content than those of the reactants or of the reaction products.

Complex reaction mechanisms can consist of a number of electron transfer steps, with some chemical steps preceding or succeeding the electron transfer steps or taking place in between them. Most organic electrochemical reactions are complex, involving large numbers of electrons in the overall reaction. Usually one step in the reaction encounters the largest energy barrier. The rate of occurrence of this step limits the rate of the overall reaction (*i.e.,* all other steps must occur at the same net rate, although they could provide for a much faster overall change). This step is called the rate-determining step and, for most practical purposes, all intermediate steps before and after it can be considered to be in equilibrium. (It is interesting to note that whenever this is the case, the Butler-Volmer equation [equation 3 below] is applicable, but with specific values of the transfer coefficients α_a and α_c characteristic of the mechanism of the reaction.)

Complex reaction mechanism

Experimental studies. Measuring the rates of electrochemical reactions (*i.e.,* current densities) as functions of electrode potential under steady-state conditions represents the normal tool of electrodics. Meaningful results could not be obtained, however, until the sensitivity of electrochemical reactions to impurities was realized and high purity techniques were introduced. Even so, the steady-state method often has shortcomings except for relatively slow electrode reactions. In many cases concentration changes at the electrodes prevent using a sufficiently wide current density range for obtaining meaningful Tafel relationships (see Butler-Volmer equation below under *Calculations*). Hence, so-called transient methods have been developed in which one electrochemical factor in the situation is rationally perturbed and the time dependence of others observed. One such method consists of placing a constant current pulse upon an electrode and measuring the variation of the resulting current through the solution. This is called the galvanostatic method for measuring the rate of an electrochemical reaction. Applying a potential pulse while observing the variation of the rate as a function of time constitutes the potentiostatic method. A third method, called the potentiodynamic, or potential sweep, method involves observations of the current as a function of the potential, while the latter is varied at a constant, known rate.

Galvanostatic, potentiostatic, and potentiodynamic methods

The advantages of transient methods over steady-state ones, in which behaviour before the attainment of the steady-state is not part of the observation, are manifold. If observations are made at sufficiently short times, events can be recorded before the onset of concentration changes, and pure activation values can be found. Hence, Tafel relationships can be obtained over a larger current density range (see below *Calculations*) than if one makes measurements over longer times, as is required in the steady-state methods. The structure of the transient states can reveal important information, such as double layer capacitance and surface coverage of the electrode by intermediate species.

Several so-called kinetic parameters, for example, partial derivatives of current density and potential with respect to concentration of chosen reactants, can be extracted from experimental measurements.

Electrochemical measurements have a limited capacity to reveal the state of an electrode surface. Nonelectrochemical methods of studying electrode surfaces, therefore, have been stressed. Optical methods have considerably gained in importance. Ellipsometry (*i.e.,* measuring changes in basic properties of polarized light as it is reflected from an electrode surface) was the first method that made possible a study of monomolecular layers of oxides and adsorbed oxygen as adsorbed organic molecules. Adaptation of such a method to transient use allows change in the surface to be related to the passivity of metals. Another kind of spectroscopy enables infrared spectra of species adsorbed at electrode surface to be taken. Mössbauer spectra may lead to an identification of thin layers on an electrode surface. A good future is seen for further development of refined techniques for the study of electrochemical processes by various combinations of spectroscopic and electrochemical means.

Types of complex reactions. *Electrocatalysis.* The problems related to the increase of rates of electrochemical reactions, or, to put it another way, the decrease of overpotential, needed to perform reactions at a given rate are the subject of electrocatalysis. Both increase and decrease are of considerable practical importance since they affect the economics of electrochemical processes. Electrocatalysis is concerned with the electrode as a substrate, or base, for electrochemical reaction and with the effect of its bulk and surface properties on the rate of reaction. Contrary to expectation, the rate of the basic step in electrochemical reaction—electron transfer—is independent of the ease with which electrons are released from the metal. Hence, those simple electrode processes in which no other changes but electron transfer take place have heats of activation virtually independent of the metal substrate.

In a reaction involving the formation of a chemical bond between the electrode substrate and one of the radicals (charged particles) formed on the surface, rates of reaction at a given potential may vary for different substrates by many orders of magnitude. The variation is a function of the strength (or energy) of bonds established between the intermediate species and the surface, and when the substrates are transition metals, regular relations between reaction rate and certain characteristics of the metal's electronic structure are observed. If such bonds are too weak, the catalytic effect of the substrate is small. If they are too strong, the intermediate species is too stable to react further, and the surface of the metal becomes virtually blocked for further reaction. Thus, there is an optimum bond strength that can be seen if the rate of a reaction is plotted against a property of the metal substrate, which is proportional to the ability of the surface to form bonds; *e.g.*, its heat of sublimation.

Electrochemical reactions can also be catalyzed or inhibited by foreign species present in the electrolytic solution. In the most general case, ions can change the rates of reactions by changing the properties of the double layer by specific adsorption effects. Organic molecules inhibit electrochemical reaction by blocking the surface. For some organic ions or dipoles, the blocking effect can begin at a certain well-defined potential. Before attaining it, the electrode reaction gives the usual current density-potential relationship, which is then suddenly interrupted by an abnormal fall of the current, corresponding to adsorption.

Electrocrystallization. Deposition of metals and other substances at electrodes as a consequence of an electrode process exhibits a number of specific features. The electrode process is followed by crystal building, and this results in a continuous change of the electrode surface. This change, in turn, affects the electrochemical properties of the system—the double-layer capacity and the rate constants of the charge transfer processes. Hence, if electrochemical properties are to be studied, transient methods should be employed that allow measurements to be made before major changes in surface morphology (structure) take place.

Since the two steps, the discharge of ions at the electrode and the incorporation of the discharged ions into the crystal lattice, are separated in time and space, an intermediate species exists at the surface, that of relatively loosely bound and freely moving atoms, called adatoms. Since the electrons tend to join the rest in the bulk of the metal, adatoms appear to have a partial charge, less than that of the elementary positive charge. The adatoms therefore attract solvent molecules, and the species is partially solvated. This reaction justifies considering an adatom as a kind of adsorbed ion, called an adion, which, however, has already undergone partial discharge.

Because such an intermediate state is possible, ions need not be reduced to the neutral atomic state at the point of incorporation into the crystal lattice. Indeed, energy considerations reveal this reduction to be improbable. Instead, discharge of ions is favoured on crystal planes. Their motion from the location of discharge to a site on the crystal occurs by surface diffusion. When such surface diffusion is inhibited, it can, in some cases, control the rate of reaction.

Problems of the crystal-building step in electrocrystallization are in many respects identical with those of crystal growth from the gas phase. Crystal building can occur in any one of three possible ways.

Kinks, naturally existing at any metal surface, form a suitable half-lattice position at which an atom is surrounded by one-half of the number of atoms that would surround it in the bulk of the metal; there, adatoms can be successively trapped and thus the crystal lattice is extended along a crystal edge and further on across the surface. This step growth mechanism is shown in Figure 11A. The mechanism, however, has a limited capacity for crystal growth. A step can move as far as the edge of a crystal, and step growth would lead to smoothing of the surface to perfection, but then further growth would cease.

Mechanisms associated with screw dislocations, or twinning edges, can provide for a continuous growth of crystals. The screw dislocation mechanism, shown in Figure 11B, is made possible by a specific fault often found in the crystal lattice that may be called a dislocation originating from a shift of one atom in the lattice with respect to a perfect arrangement. This shift may then result in the formation of a monoatomic edge projecting above the electrode surface at which new atoms can be stacked. The stacking produces a turning of the edge around the base atom as a centre, the process producing a spiral growth.

Stacking

If growth sites are rare, or if the substrate at which deposition should take place is foreign to the depositing metal, the charge transfer results in an accumulation of adatoms to a concentration considerably larger than that which can exist there at equilibrium with the crystal lattice. This condition is termed supersaturation. In such a situation, agglomeration of adatoms to form crystal nuclei is favoured. Surface energy requirements show that at any degree of supersaturation, nuclei of certain dimensions are stable and can represent sites for further growth. The process is shown in Figure 11C.

The formation of deposits is controlled by the above mechanism as long as the discharge process supplies ample amounts of adatoms as building material over the entire surface. If the rate of deposition is increased, so as to produce near the surface considerable depletion of ions, uneven deposition will start. This is caused by the protrusions of a normally rough metal surface being closer to the bulk of solution than the recessed parts and, hence, getting

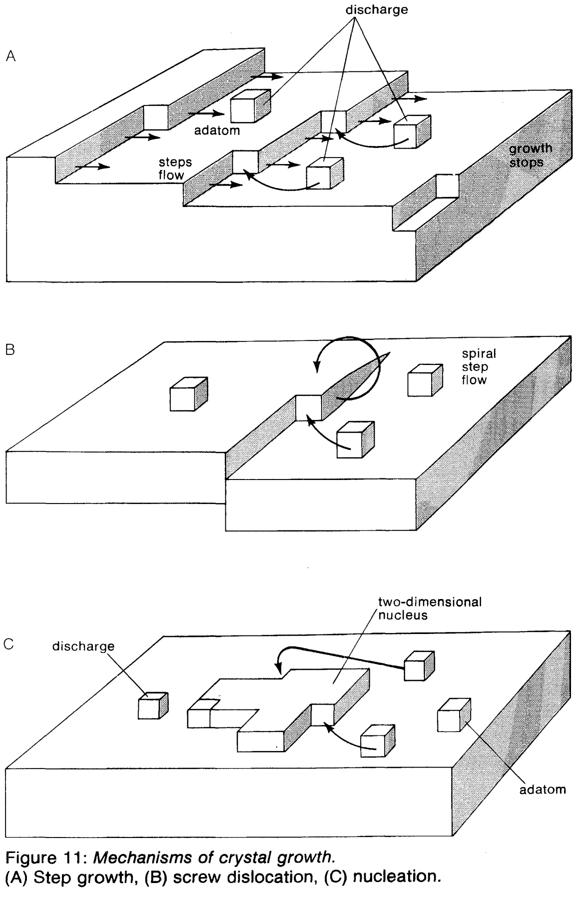


Figure 11: *Mechanisms of crystal growth.*
(A) Step growth, (B) screw dislocation, (C) nucleation.

a somewhat faster supply of the discharging species. Once such a situation is established, it tends to develop further. The faster growing points penetrate into ever richer layers of solution, resulting in ever faster growth. Thus, a natural consequence of deposition under transport-controlled rate is the amplification of the original surface roughness and the appearance of protruding spikes, called dendrites, as deposits.

A converse process is the smoothing of the original surface irregularities, which may occur when some foreign species, called an additive, is present in solution and adsorbs on the surface and inhibits the process of discharge. If those molecules are incorporated into a growing deposit, a situation may arise in which their supply to recessed parts of the surface becomes slower than to elevated parts. As a result, deposition becomes faster at recessed parts than at elevated ones and leveling of the surface occurs. This process has considerable technical application. Metal dissolution can sometimes be governed by a similar, transport-controlled mechanism, in which case a polishing effect is obtained. Electropolishing has found wide application in practice.

Organic electrode reactions. A very large number of electrochemical reactions involving organic molecules are known. An example is the oxidation of ethylene according to the equation:

$$\underset{\text{(ethylene)}}{C_2H_4} \quad + \quad \underset{\text{(water)}}{4H_2O} \quad \rightarrow \quad \underset{\substack{\text{(carbon}\\\text{dioxide)}}}{2CO_3} \quad + \quad \underset{\substack{\text{(hydrogen)}\\\text{ion)}}}{12H^+} \quad + \quad \underset{\text{(electrons)}}{12e^-}.$$

Many chemical organic reactions can be made to function electrochemically, a general advantage being that the rate of the reaction is easily controlled by controlling the potential. A change of potential may change the path of the reaction and hence a certain product may be tuned in. In particular, polymerization reactions at electrodes can be stimulated; such a method is used as a step in the production of nylon.

Ductile and brittle properties

Electromechanical effects. Several properties of solid bulk metals turn out to be dependent upon electrochemical reactions. Thus, if the hydrogen atom concentration on an electrode surface exceeds a certain critical value, the hydrogen molecules (H_2) inside intrinsic voids in the metal's interior may exceed a critical gas pressure and cause the voids to open up and spread, thus causing the metal to lose ductility and become brittle. It is the electrochemical reaction, represented by the equation $H^+ + e^- \rightarrow H$ (hydrogen ion plus electron equals hydrogen atom) at the metal–solution interface that controls the surface hydrogen atom (H) concentration and, hence, the hydrogen molecules (H_2) present inside the metal. A limiting electrode potential which causes hydrogen atom concentration exists, above which a metal loses strength. This concept applies also to moist metal surfaces; *i.e.,* most practical metal surfaces, the electrolyte being in effect a surface moisture layer and the ions being produced from the dissolution of atmospheric carbon dioxide. Friction between solids involving moist or wet surfaces is also dependent on electrochemical factors.

Multielectrode systems. So far, systems have been considered in which a single electrode process takes place. In principle, at any electrode potential all species present in the system fall into two categories: those that are stable, and those that undergo oxidation or reduction. The stable species are those that at the given electrode potential would not decrease their free energy by giving off or accepting electrons.

At any potential there should occur a codischarge of all unstable species. Thus, the system can be considered as a multielectrode system, consisting of as many electrodes as there are redox couples present. The rate at which different processes occur, however, can be so widely different that usually a single process is by far the dominant one. Systems at which two electrode processes occur at comparable rates are of considerable importance. If two kinds of metal ions are discharged simultaneously, an alloy is formed upon crystallization. The properties of the alloy would in most cases be those determined by the phase diagram (a plot of the temperature of melting versus the composition of the mixed system) for the ratio of quantities of discharged metals given by the rates at which the discharges take place. In many cases, nonequilibrium metal phases are formed giving unusual properties to the alloy.

Corrosion

If a process of metal dissolution (an oxidation process) can occur at a rate comparable to that of some reduction process on the same metal, a corrosion couple is established. Thus, zinc immersed in acid solution tends to establish a potential sufficiently positive for the metal to eject metal ions into the solution (oxidation). At the same time, it is also sufficiently negative for the reduction of hydrogen ions, present in any aqueous solution, to hydrogen gas. Hence, a spontaneous process of hydrogen evolution and of dissolution of the metal will take place. The rate at which this corrosion process occurs is governed by rate laws of the type given by equation (3) (see below *Calculations*). The mixed potential, spontaneously established by the corroding metal, is obtained by equating expressions for the anodic and cathodic currents of the two processes in the corrosion couple.

Corrosion can be prevented in two ways. One is by using an external source to make the potential of the metal sufficiently negative to bring it into the potential region in which the metal is stable, called cathodic protection; and the other is by provoking by some means the formation of a film on the surface that would slow the process. Such films could consist of an oxide or a layer of organic molecules that prevents dissolution and hence is called an inhibitor.

APPLICATIONS

Electrochemical processes are used in many ways and their use is likely to increase because they can replace polluting chemical situations with nonpolluting electrochemical ones. In many fields, however, applications have been profitable for some time. Major categories are listed below.

Metallurgy. All technologically important metals, except iron and steel, are either obtained or refined by electrochemical processes; for example, aluminum, titanium, alkaline earth, and alkali metals are obtained by electrodeposition from molten salts, and copper is refined by electrolysis in aqueous copper sulfate solutions.

Electroplating. One of the major ways of both decorating objects and improving their resistance to corrosion is by electroplating them. All major metal-working industries, particularly the automobile industry, have large electroplating plants.

Chemical industry. Electrolysis of brine to obtain chlorine and caustic soda is an electrochemical process that has become one of the largest volume productions in the chemical industry. Modern processes cover a wide field, from the production of a variety of inorganic compounds to the production of such synthetic fibres as nylon. Intensive research in organic electrochemistry promises major developments in application, particularly with the prospect of greatly reduced electricity costs expected eventually to arise from the development of controlled fusion.

Batteries. Electrochemical storage of electricity is effected in batteries. Such devices are electrochemical cells and consist of two electrodes per unit. As the electricity to be stored is accepted on the plates of the cell, it converts substances on the plates to new substances having a higher energy than the old ones. When it is desired to make the electricity available again, the terminals of the battery are connected to the load and the substances on the battery plates retransform themselves to those originally present, giving off electricity as a product of their electrochemical reactions. The steadily rising production of the lead-acid battery is largely the result of its use for starting the internal-combustion engine, which has had an equally steady rise. Other electrochemical systems are also used as storers. The nickel-iron (Edison cell) and nickel-cadmium battery with alkaline electrolyte are both used in applications where longer lives than those of the lead-acid battery are needed; the silver-zinc battery is used to start airplane engines because of its high power per unit of weight. A variety of new systems is being investigated for covering other needs. One of the greatest challenges to electrochemists and electrochemical engineers is that of producing a battery with sufficient power and energy density to run an automobile the way gasoline (petrol) does. Even if the best hypothetical predictions for removal of polluting chemicals from automobile exhausts is realized, the cleanup will not be sufficient because the expected growth of the automobile population will continue to increase the pollutant rate.

Fuel cells. The energy of chemical reactions is converted into electrical energy in fuel cells. In these, the fuel (*e.g.*, hydrogen, hydrazine) is fed continuously to one electrode, while oxygen from the air is reacting at the other one. The efficiency of energy conversion in fuel cells is more than twice that attainable by conventional means, for example, by means of internal combustion.

Analytical chemistry. In analytical chemistry, most modern automated instrumental analysis is based on electrode processes—for example, potentiometry, used to measure ionization constant.

Biological research. In biology the idea that many biological processes, from blood clotting to the transfer of nerve impulses, are electrochemical in nature continues to spread. The biological conversion of the chemical energy of food to mechanical energy takes place at an efficiency so high that it is difficult to explain without electrochemical mechanisms. Intensive research is developing in various directions in bioelectrochemistry.

CALCULATIONS

The three equations referred to above are stated in this section, and other mathematical considerations are also included. The rate of an electrochemical reaction in terms of oxidation and reduction reactions, the concentration of the reacting species, the electrode potentials and the current densities can all be related quantitatively according to equation (1):

$$i = \overrightarrow{i} - \overleftarrow{i} = \overrightarrow{k} C_{\mathrm{red}} \exp\left(\frac{\alpha_a F}{RT} E\right) - \overleftarrow{k} C_{\mathrm{ox}} \exp\left(\frac{\alpha_c F}{RT} E\right)$$

in which i is the net current density, $\overrightarrow{i}$ and $\overleftarrow{i}$ are the partial current densities of the oxidation and reduction respectively, C_{red} and C_{ox} are the concentrations of the reducing and oxidizing agents, respectively, $\overrightarrow{k}$ and $\overleftarrow{k}$ are the corresponding rate constants, while α_a and α_c are the so-called transfer coefficients; *i.e.*, specific constants giving a proper influence factor to the exponential dependence of rate on the potential, E. In the case of a simple one-electron transfer these factors are termed symmetry factors, for they, in a way, reflect the symmetry of the energy barrier. It can be proved that $\alpha_c + \alpha_a = n$, the number of electrons exchanged in a single act of an electrode reaction.

Nernst Equation. For a particular value of E the two partial current densities must become equal. This value of potential is the reversible electrode potential. From equation (1) one can deduce equation (2):

$$E_{\mathrm{rev}} = \frac{RT}{(\alpha_c + \alpha_a)F} \ln \frac{\overrightarrow{k}}{\overleftarrow{k}} + \frac{RT}{(\alpha_c + \alpha_a)F} \ln \frac{C_{\mathrm{ox}}}{C_{\mathrm{red}}}$$

$$= E^\circ + \frac{RT}{nF} \ln \frac{C_{\mathrm{ox}}}{C_{\mathrm{red}}}.$$

This equation is known as the Nernst Equation; E° is the standard electrode potential (at $C_{\mathrm{ox}} = C_{\mathrm{red}} = 1$) characteristic of the given redox couple.

The standard electrode potential on the hydrogen scale is related to the thermodynamics of the electrode process. It reflects the standard free energy change of the redox reaction between the electron and the given redox couple, relative to the free energy change that takes place in the hydrogen electrode process.

Butler-Volmer Equation. The reversible electrode potential can be introduced into equation (1) and the potentials taken relative to its value. When so expressed, they are termed overpotentials and can be stated as $\eta = E - E_{\mathrm{rev}}$; equation (1) then transforms to equation (3):

$$i = i_0 \left\{ \frac{(C_{\mathrm{red}})_i}{(C_{\mathrm{red}})_o} \exp\left(\frac{\alpha_a F \eta}{RT}\right) - \frac{(C_{\mathrm{ox}})_i}{(C_{\mathrm{ox}})_o} \exp\left(-\frac{\alpha_c F \eta}{RT}\right) \right\}$$

in which i_0 represents the value of either of the (equal) electron-emitting and electron-accepting partial current densities at the reversible potential and is termed the exchange current density. Equation (3) is called the Butler-Volmer Equation and represents one of the most fundamental relationships of electrochemistry.

As overpotentials, either positive or negative, become larger than about 5×10^{-2} volts (V), the second or the first term of equation (3) becomes negligible, respectively. Hence, simple exponential relationships between current (*i.e.*, rate) and overpotential are obtained, or the overpotential can be considered as logarithmically dependent on the current density. This theoretical result is in agreement with the experimental findings of the German physical chemist Julius Tafel (1905), and the usual plots of overpotential versus log current density are known as Tafel lines. The slope of a Tafel plot reveals the value of the transfer coefficient α for the given direction of the electrode reaction.

Difficulties in transport of reaction species. The above conclusions about the overpotential-current density relationship are valid as long as the ratios of concentrations at the electrode surface of the species involved at current density i, C_i and in the absence of current C_o stay close to unity. As the current density is increased, the concentration gradient needed to maintain a corresponding diffusion flux of the species concerned must begin to become appreciable. This condition is possible only if the concentration of the species at the surface starts to differ appreciably from the bulk value; *i.e.*, $(C_i)_i/(C_i)_o \neq 1$. The change in concentration of the discharging species at the electrode surface with time can, in principle, be obtained by using a second order partial differential equation (Fick's law), which, however, has explicit solutions only for a limited number of well-defined boundary conditions.

When significant concentration changes set in, no more exponential dependence of current density on potential can be obtained. It can be derived that, instead, a transition toward a limiting value takes place.

The important case is that in which the concentration of the discharging species at the electrode surface becomes equal to zero. The steady-state (*i.e.*, independent of time) current density obtained in such a case is the highest possible for the given set of conditions (diffusion limiting current density). The value of the concentration gradient in this case is directly proportional to the bulk concentration of the species involved and inversely proportional to the thickness of the diffusion layer (*i.e.*, the layer close to the electrode in which the concentration of the species differs from the species concentration in the bulk). This layer most often has a thickness fixed by hydrodynamic conditions in the solution surrounding the electrode. The definition used most often for the diffusion layer thickness is that of the German physical chemist Walther Hermann Nernst (1864–1941), according to whom this quantity is equal to the distance from the electrode at which the concentration would reach the bulk value if the concentration gradient were constant and equal to that at the electrode surface.

If a current larger than the limiting current is forced upon the electrode, the given electrode process will be able to sustain it only in the initial period in which the layer of solution close to the electrode is not completely exhausted of the discharging ions. As the concentration of ions tends toward zero, the electrode potential will change and another electrode process will start. The time at which the abrupt change of potential toward a new process takes place is termed the transition time. The relationship between transition time, current density, and concentration of the discharging species is given by Sand's equation:

$$\tau^{1/2} = \frac{\pi^{1/2} n F C_o D^{1/2}}{2i}.$$

Since τ is a well defined function of the concentration of the discharging species, the observation of the transition time can also be used as an analytical tool (chronopotentiometry). (J.O'M.B./A.De.)

Photochemical reactions

Photochemical reactions are chemical processes initiated by the absorption of energy in the form of visible, infrared, or ultraviolet radiation. The immediate consequence of this absorption is called the primary photochemical process. Subsequent changes, called secondary processes, are part of photochemistry. The process by which a photochemical reaction is carried on is photolysis.

Photochemistry differs from most other aspects of chemistry in one regard. If an atom or molecule absorbs energy from a beam of light, it gains far more energy than it ever could by other methods; *e.g.,* from ordinary heating. Consequently, photochemical processes are sometimes extremely efficient for the conversion of energy from light into chemical energy. The most important single photochemical process for living systems, photosynthesis, is of this type: green plants convert light energy into stored chemical energy by producing carbohydrates from carbon dioxide and water. Many other natural processes are wholly or partly photochemical. Ozone in the upper atmosphere, which shuts out most of the Sun's intense ultraviolet radiation, is produced by the action of light on oxygen molecules: sunlight breaks some of the molecules into atoms, which combine with other oxygen molecules to make ozone.

Natural photochemical processes

Bleaching laundry in sunlight is at least partly a photochemical process, and so is the darkening of white lithopone (zinc sulfide) paint. Photography is based on a photochemical process, the action of light on grains of silver chloride or silver bromide.

GENERAL CONSIDERATIONS

Kinds of photochemical effects. The action of light on a chemical system may take various forms:

1. Light may act as a booster to maintain a reaction that might otherwise proceed at an imperceptibly slow rate, the reacting molecules exchanging little or no net energy with their environment. Some molecular rearrangements (reactions that alter the geometry of individual molecules) fall into this category.

2. Light may act as a trigger, initiating a reaction that, once started, can proceed spontaneously. An example is the initiation of the reaction between hydrogen (formula, H_2) and chlorine (Cl_2) to give hydrochloric acid (HCl). Hydrogen and chlorine can be kept mixed together for an indefinite period at room temperature in a dark container; but if violet or ultraviolet light reaches the mixture, the two substances react explosively, giving off considerable stored chemical energy as heat. The primary photochemical process in this case is the formation of two chlorine atoms from a chlorine molecule, Cl_2.

3. In some photochemical processes, absorbed light converts the reacting molecules to a new state, higher in energy than their initial state; in this case the energy of the absorbed light is converted into chemical potential energy and is stored that way. Most photochemical processes occurring in biological systems are of this type; *e.g.,* ultraviolet irradiation converts ergosterol, found in plant and animal tissues, into biologically active vitamin D.

4. A photochemical system may store light energy from the Sun as chemical energy and then, only incidentally to the photochemical process, release that energy as electrical energy. This reaction occurs in the solar battery, used to supply the power in satellites and space vehicles.

All these processes have one common characteristic: each, at some stage, requires that a large quantity of energy be available to a single reacting atom or molecule. The requirement may be a thermodynamic one; that is, that the products simply contain more stored energy than the reactants. Alternatively, the requirements may be kinetic; in this case the products may store more or less energy than the reactants, but some step along the reaction path requires a large quantity of energy to surmount a barrier, as it were. (The hydrogen–chlorine reaction has as its barrier the requirement that a chlorine–chlorine bond be broken to convert a chlorine molecule into its two constituent atoms.) In more complex molecules, the rearrangement from one shape to another exhibits an energy barrier because, between the initial and final structures, there must be a series of intermediate shapes of higher potential energy than either the initial or final shapes.

The theoretical aspects of photochemistry are derived largely from the principles of atomic and molecular physics and of quantum mechanics, while its analytical methods are mainly those of spectroscopy.

The development of photochemistry. The hydrogen–chlorine reaction cited above, perhaps the first recognized photochemical reaction, was first observed about 1801 and studied in detail some 40 years later. That only the absorbed light can affect a chemical reaction was recognized early, but this qualitative idea could not be stated as a quantitative law until the development of the modern theory by Einstein. Early photochemistry was largely confined to the identification of reaction products, often resulting from the action of sunlight.

Einstein's work in light absorption

The basic process by which light is absorbed by matter was first proposed by Einstein (1905, 1912, 1916). His work was directed particularly toward the photoelectric effect, in which the primary result of the absorption of radiation by a metal is the emission of free electrons; but the interpretation is general and provides the foundation of modern photochemistry: a beam of light is a beam of photons, or quanta (energy packets), of electromagnetic radiation (see below *Photochemical quantities*). The energy of each photon is proportional to the frequency of the light, and the total number of photons is proportional to the intensity of the light. The primary photochemical process then follows directly from the absorption of one photon by one atom or molecule. (If the radiation field is very intense, the concentration of photons may be so high that some molecules absorb two photons.)

After 1910 a German physicist, Emil Warburg, began a detailed study of the formation of ozone (O_3) by the action of ultraviolet light on the normal oxygen molecule (O_2). This work represented the first measurement of the number of product molecules formed for each quantum of energy absorbed: the quantum yield of the reaction. In the next few years most of the foundations of modern photochemistry were completed.

The role of chain reactions in photoprocesses was first elucidated for the hydrogen–chlorine reaction in the period 1913–25, and photochemistry grew steadily in the next three decades. From about 1955 the field expanded rapidly as it came to be used as a tool for the synthesis of organic molecules and was extended to the study of very short wavelengths (including X-rays and gamma rays) and of chemical phenomena produced by the optical amplifier, or laser, the acronym for light amplification by stimulated emission of radiation.

THE PHOTOCHEMICAL PROCESS

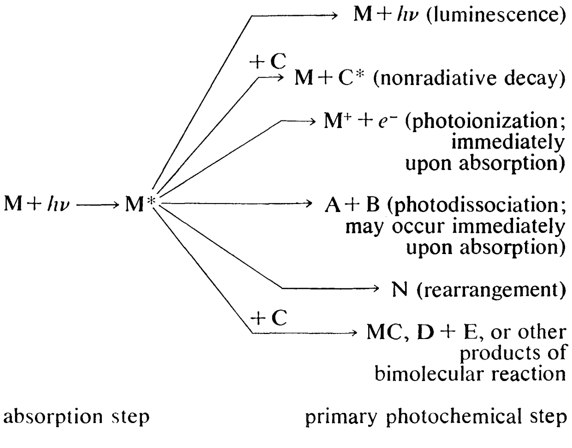
Photochemical quantities. The most fundamental ideas and the basis of all quantitative interpretations in photochemistry are the quantum-mechanical principles that all energy exists in packets, called quanta, and that the quanta of electromagnetic radiation, called photons, act like particles. The energy of each photon is directly proportional to the frequency of light—a relationship that is expressed by the equation $E = h\nu$. In this equation, E is the energy in ergs, and the frequency in cycles per second is symbolized by the Greek letter nu, ν. The constant of proportionality, h, is called, after its discoverer the German physicist Max Planck, Planck's constant; it has the value 6.625×10^{-27} erg-second. (One erg equals the work required to cause a mass of one gram to accelerate at one centimetre per second per second through one centimetre.) Alternatively, since the frequency, ν, and the wavelength, symbolized by the Greek letter lambda, λ, of light are related to c, the velocity of light (3×10^{10} centimetres per second), by the equation $\nu\lambda = c$, the energy of a quantum can be related to the wavelength of the corresponding radiation rather than to its frequency. This relationship is expressed by the equation $E = hc/\lambda$. When E is expressed in ergs and λ in angstrom units, the proportionality becomes $2 \times 10^{-8}/\lambda$; when E is in electron volts per photon, however, it becomes $12{,}400/\lambda$; and when E is in kilocalories (kcal) per mole of photons it is $2.85 \times 10^{5}/\lambda$. The angstrom unit (Å) is 10^{-8} centimetre and is commonly used to designate wavelength. The electron volt (eV), 1.602×10^{-12} erg, is the usual unit for measuring energies of individual particles in atomic physics.

By means of these quantities the relationship between ordinary chemistry and photochemistry can be expressed more precisely in terms of some simple magnitudes. Green light, for example, has a wavelength of about 5000 angstroms. (The entire visible region lies between about 4000 and 7000 angstroms.) A photon of green light, consequently, has an energy of about 2.5 electron volts, or an einstein (one mole of photons) of green-light photons has an energy of 57 kilocalories. Ultraviolet light of wavelength 2500 angstroms has just twice as much energy per photon or per mole. By comparison, a weak chemical bond requires an energy of about one electron volt to be broken, while a very strong bond requires more than 10 electron volts. In contrast, however, two molecules colliding at room temperature strike each other with an energy of only about 0.05 electron volt.

Types of energy in absorbers

Absorption of electromagnetic energy. The history of a photochemical reaction necessarily begins with the absorption of a quantum of electromagnetic energy by an atom or molecule. Initially, the absorbing molecule or atom is in a state in which its energy is relatively low. The energy contained in the absorber is essentially of four types: (1) kinetic energy of motion through space (translational energy); (2) energy of rotational motion about its centre of mass; (3) energy of vibration of the component atoms against each other if the species is a molecule; and (4) energy of the negatively charged electrons moving in the field of force produced by the other electrons and the positively charged nuclei. The energy change initiating a photochemical process virtually always involves excitation of one electron from a state of low energy to a state of higher energy. When the absorber is a molecule, changes in rotational and vibrational energy usually also occur; the translational energy of an atom or molecule, however, is almost completely unaffected when a quantum of radiation is absorbed, except in the case of highly energetic gamma radiation.

Usually, in the initial low-energy state of an absorber, the electrons are in their lowest possible state of energy; that is, their lowest quantum state. The electrons cannot absorb all incident radiation, but only those quanta with energies corresponding to the allowed quantum jumps for that particular absorber. It has been determined that a hydrogen atom in its lowest state can absorb only those quanta having energies that fit the expression $13.7\,(n^2 - 1)/n^2$eV, in which n may equal 2, 3, . . . ; according to this condition, it cannot absorb a six-electron-volt photon, for example. The energies that quanta must have in order to be absorbed by a molecule or atom correspond to the characteristic frequencies of the absorption spectrum of that substance.



Primary photochemical step. The occurrence of a photochemical reaction, then, begins with the absorption of a light quantum of suitable energy. The energy is used mainly or entirely to raise one electron into an excited state, with the result that the molecule or atom itself is in an activated state. The primary photochemical step then occurs; it may take the form of any of a number of different processes, as shown schematically below, in which M is the absorber; A and B indicate products formed by dissociation of M; C is a second substance, which interacts with M; N is a rearrangement product of M; D + E are secondary products that result from the interaction of M and C; e^- is an electron; and the asterisk designates electronically excited substances.

Photo-ionization and photo-dissociation

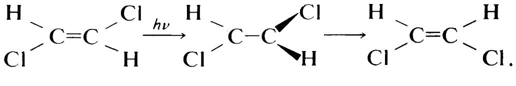
For each kind of molecule, more than one kind of excitation usually is possible. One of the origins of continuous absorption is photo-ionization, in which case the excited state of the electron is one in which it is free to escape from the atom or molecule, leaving behind a charged particle, or ion. It is the one case in which absorption and primary reaction are essentially identical. Photo-ionization can be thought of as a prototype—perhaps the most elementary photochemical process. Continuous absorption is characteristic of a transition in which a new free particle is produced: it may be an electron, as in photo-ionization, or it may be an atom or a molecular fragment. These latter two classes of particle are produced by photodissociation, a second type of photochemical process, which usually occurs when the electron being excited moves away from a locale where it can participate in a chemical bond into some other region where it is unable to use its attraction to pull two nuclei together. In normal photodissociation the excited electron remains bound to one or the other of the products of the photolysis. Photodissociation is sometimes almost as directly linked to the absorption process as photo-ionization is, but it can also be considerably slower.

Absorption in a line or band spectrum leaves all the particles bound together with their geometrical configuration essentially unchanged, at least for a short time (10^{-14} second). The change produced by such an absorption involves the excitation of one electron from a closely held low-energy state to a more loosely bound excited state. In many cases there are two or more possible excited states and several electrons that could undergo a transition; because each spectral line or band corresponds to one specific electronic transition, however, it is frequently possible for an experimenter to study one kind of electron excitation at a time. This condition can be achieved by irradiating the sample with light lying only in the energy range corresponding to the transition that the experimenter is interested in. An experiment of this type can often be done with light of almost a single pure frequency (monochromatic light).

Luminescence and nonradiative decay

As soon as an electron is excited and the molecule is in an activated state, several events can occur. The molecule may return to its initial state, without reacting, in either of two ways: it may radiate, giving up a light quantum by luminescence; or it may lose its excitation energy, without radiating or reacting, by undergoing nonradiative decay, a process that requires effective collision partners to carry off the excitation energy. Nonradiative decay usually goes through two stages. In the first step, little or no energy is lost; the energy of electronic excitation is simply converted into vibrational energy (called intersystem crossing, or internal conversion). In the second stage, the excess energy is given up to the environment gradually through a succession of collisions.

If the excited molecule does react, it may do so in any of several ways. It may undergo photo-ionization or photodissociation, as indicated above. Or it may undergo an intramolecular reaction (an internal rearrangement) or an intermolecular process. An example of intramolecular rearrangement is the conversion of *trans*-1,2-dichloroethylene into its *cis* isomer (a molecule with identical numbers and types of atoms but different arrangement of them). In the equation for this reaction, the lines represent single covalent bonds and the shaded lines provide perspective; *i.e.*, their pointed ends are away from the viewer, their blunt ends toward the viewer:



The product, in this case, has almost the same energy as the starting compound; the excitation energy required to reach the twisted intermediate form must be given off somehow by the product, usually by collision.

Primary intermolecular reactions may involve nothing more than the transfer of a quantum of energy, as with mercury (Hg) and ethylene (C_2H_4):

$$\text{Hg (vapour)} + h\nu \longrightarrow \text{Hg}^*$$

$$\text{Hg}^* + C_2H_4 \longrightarrow \text{Hg} + C_2H_4^*,$$

leaving the excited ethylene to undergo secondary reaction. The intermediate kind of primary nondissociative reaction uses collisions in a way akin to the nonradiative decay process. An excited molecule undergoes a collision that converts it into a new reactive state without removing much of its excitation energy. The collision may change the geometry of the reactive molecule into a form that can rearrange bonds or dissociate readily, or it may change the electronic state of the molecule through internal conversion or intersystem crossing. Changes in electronic state usually involve only the single excited electron, but sometimes the molecular geometry is also involved.

In order for any state, be it the initial excited state or any subsequent one, to play a part in a photochemical reaction, it must endure long enough for the reaction to occur. Each photochemical reaction competes with radiative and nonradiative de-excitation to consume the activated molecules. A few typical numerical examples will illustrate the way this competition goes on. Fast luminescence, called fluorescence, occurs with a half life of about 10^{-8} second; *i.e.*, roughly half of the excited molecules in any sample will radiate their excitation energy away in about 10^{-8} second provided they have an allowed path for fluorescence and provided they are not destroyed faster by some other process. Simple dissociations of excited molecules occur much faster: an excited molecule in an unstable state dissociates in about the time it takes an atom to go through one vibrational period, or about 10^{-13} second. Consequently, a molecule excited to a truly dissociating state has virtually no chance to radiate and will indeed follow the path of dissociation. Other unimolecular processes may occur with characteristic times ranging from 10^{-13} second up to a few seconds, if no other process interferes.

Fluorescence and phosphorescence

Radiative processes can also be slow. Slow luminescence is called phosphorescence and is the result of emission of a quantum from a molecule undergoing a relatively improbable transition. Fluorescence is the result of an "allowed" transition. Phosphorescence frequently is characterized by a lifetime for the excited molecule between 10^{-5} and 10^{-2} second, but many cases outside these bounds are known. Fluorescence, phosphorescence, dissociation, and some rearrangements involve only a single molecule and have specific rates (rate per unit of reactant concentration) or lifetimes for reactant molecules that are independent of the number of molecules present. Processes involving two molecules have specific rates that depend on the concentration of collision partners, because these rates depend on the frequency of collisions. The maximum rate for a bimolecular reaction is just the collision rate; this rate, about 10^{29} per cubic centimetre per second in the normal atmosphere, is attained only if the reaction occurs with every collision.

In solution, the maximum bimolecular rate is governed by the rate at which the reactive species can diffuse together through the liquid or solid medium. Like the collision frequency, the maximum encounter rate represents an upper limit for the rate of reaction. Rates in solution are inversely proportional to the solvent viscosity, with the result that bimolecular processes generally proceed slower in glycerin, for example, than in water. Bimolecular de-excitation processes are usually limited by the rate at which the internal electronic state can change.

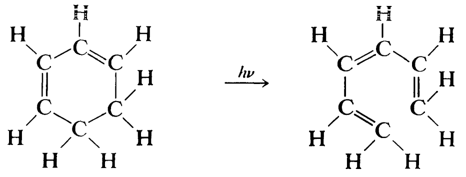
Any molecule that may decay or react in more than one way will follow the fastest of these ways. Comparison of the reaction rates shows that ionization or dissociation usually is fastest, if either can occur; fluorescence is fast in gases but is slow relative to collisions, nonradiative decay, and bimolecular reactions in solutions.

When a molecule changes its electronic state without direct absorption of light, as in nonradiative decay or internal conversion, it may go into a new excited state, which, left alone, would be short-lived (less than about 10^{-6} second); or it may go into a long-lived (metastable) excited state. Both types of reaction are known in photochemistry, but the long-lived states are particularly important in storing energy for chemical processes. The most important of the last named kind of transition is a singlet–triplet transition—singlets and triplets being electronic states that differ in having paired and unpaired electrons, respectively.

Triplet state

Triplets are the most commonly found phosphorescing states. They cannot radiate rapidly and are therefore among the most long-lived excited species. For this reason triplets are frequently the primary products of photochemical reactions. Excited singlets are sometimes found, but they more often decay, often to triplets. Triplets can be identified by their magnetic properties; by contrast with singlets, the individual electron magnets of a triplet do not cancel each other so that triplets exhibit weak magnetism (paramagnetism), which can be detected by suitable techniques.

An interesting case in which the entire photochemical process occurs in the primary reaction is that of valence tautomerization; *i.e.,* the inter-conversion of molecular structures that differ very little in the relative positions of their atoms but have different arrangements of chemical bonds, as in the photoisomerization of 1,3-cyclohexadiene shown in the structural formulas below:

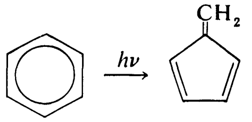

A B

Though structures A and B are nearly identical geometrically, they represent different states of the electrons.

Secondary steps. If an excited molecule does not return by a direct path to its lowest energy state, then it must dissociate, react with neighbours or itself, or revert to a new electronic state, still having some excitation energy. These are the primary steps (above). Any subsequent step is a secondary reaction. An example is the primary dissociation of an oxygen molecule (O_2) into two oxygen atoms; the only secondary reaction of importance in pure oxygen is combination of oxygen atoms with oxygen molecules to produce ozone (O_3). Both ozone and oxygen can then go on to attack hydrocarbons, in typical urban atmospheres.

Rearrangement reactions

Reactions in which atoms move from one site to another within a molecule—that is, rearrangement reactions—may be interpreted as secondary reactions following the photoexcitation process. Benzene irradiated with light between 2000 and 3000 angstroms rearranges (in low quantum yield) to the unstable hydrocarbon fulvene with the movement of a hydrogen atom, as shown in the structural formulas below:


benzene fulvene

in which the circle within the hexagon represents six electrons of benzene; the first step involves only the excitation of one of these six relatively free electrons, which do not participate in carbon–hydrogen bonds to any significant degree. Thus, only in a secondary step following photoexcitation can the atoms rearrange to the molecular structure of the product.

Frequently, the primary photochemical process is used as a tool to permit the secondary reactions to be studied. Photosensitized reactions often fall into this category. Mercury atoms can absorb light from a mercury lamp very efficiently and then, in their excited state, collide with other molecules and transfer their excitation to them. This technique has been used to dissociate nitrous oxide (N_2O) and provide free oxygen atoms. The reaction sequence proceeds as shown below:

$$Hg + h\nu \longrightarrow Hg^*$$

$$Hg^* + N_2O \longrightarrow Hg + N_2 + O.$$

(The alternative mode of decomposition, to give NO + N, has not been found to occur.) Then the oxygen atoms may react with whatever other substances are present. Measuring the amount of nitrogen produced in such a photolysis gives directly the number of oxygen atoms that were available for reaction.

Chain reactions

If a primary step involves dissociation of a molecule into two radicals (molecular fragments with single or unpaired electrons), the secondary steps are often chain reactions. About 1923 it was recognized that there were certain cyclic processes, called chain reactions, in which reaction of an active intermediate with a molecule of starting material to give a product necessarily also produced a new active intermediate. The hydrogen–chlorine reaction is an example. Light dissociates a few chlorine molecules into atoms, which then attack hydrogen molecules to give hydrogen chloride product and hydrogen atoms. The hydrogen atoms react with chlorine molecules to produce more hydrogen chloride and chlorine atoms. The reactions are:

$$h\nu + Cl_2 \longrightarrow 2Cl$$

$$Cl + H_2 \longrightarrow HCl + H$$

$$H + Cl_2 \longrightarrow HCl + Cl.$$

Reactions of this kind require an initiator (light, in this case) and a propagating mechanism and may involve a separate terminating step. Termination actually occurs either when the concentration of at least one reactant becomes very low or when another substance (*e.g.,* oxygen or iodine) is present to react more rapidly with the radicals than can the hydrogen or chlorine molecules. The reaction must also be exothermic (heat-releasing) or nearly thermoneutral (that is, neither releasing nor absorbing heat), otherwise it may cool the system so that it no longer reacts. In the reaction just cited, considerable heat is generated—usually much faster than it is radiated or transferred to the vessel walls—and an explosion is the result. An even more unstable situation—violently explosive—arises if one radical initiator produces more than one new radical.

Chain reactions always produce many product molecules for every quantum of energy absorbed; *i.e.,* the quantum yield may be much greater than one for the reaction. Reactions involving no chain processes necessarily have quantum yields near or below one, while a chain reaction may have a quantum yield of 80,000. Even a rough measurement of quantum yield will generally show whether a reaction proceeds through a chain mechanism.

EXPERIMENTAL METHODS

The methods of studying photochemical reactions have been developed chiefly to deal with problems of three kinds, which are: (1) getting the desired light energy into the system being studied; (2) determining the progress of the reaction—getting immediate pertinent information from the experiment; and (3) determining, from the results of individual experiments, the broader inferences for which the study is carried out. Photochemists commonly try to determine from a series of individual reactions the mechanism of a reaction—*i.e.,* the behaviour of each atom throughout the progress of the reaction.

Illuminating the sample. Providing suitable illumination for photochemical reactions, the experimenter has four controls: (1) The frequency, or energy, of the light can be controlled. (2) The intensity of the light can be varied to control the rate at which photons arrive at the sample and thereby to control the number of excited molecules present at one time. (3) The time of illumination can be varied to control the total number of excited molecules produced. (4) The polarization of the light may be selected.

The Sun as light source

In the earliest photochemical observations the Sun was the source. No control could be exerted, of course, on the frequency of light coming from the source itself, but simple controls of the light reaching the sample were available through the use of sample containers of differing composition; *e.g.,* flint glass transmits very little ultraviolet radiation (of wavelengths less than about 3600 angstroms), borosilicate (Pyrex) glass transmits light of wavelengths above about 3000 angstroms, and some fused silica is transparent to wavelengths shorter than 1700 angstroms. The next simplest way to control the wavelength reaching the sample is to interpose a suitable filter. Filters may transmit broad bands of frequencies or narrow and nearly monochromatic bands. A more complex step is the use of the monochromator, a device (usually a prism or grating) that permits the arbitrary choice of any frequency.

Incandescent sources, such as the Sun and ordinary electric tungsten-filament lamps, emit light of all frequencies, the total intensity and the distribution of intensity depending primarily on the source temperature. For higher intensities it is often possible to use the emission spectrum of an atomic or molecular vapour. Some common sources use mercury, cadmium, sodium, xenon, hydrogen, or deuterium in anelectric-arc discharge to provide intense and steady fluxes of radiant energy. If the lamp pressures are low, the light emitted by mercury, cadmium, or xenon is restricted to a few, intense, well-separated wavelengths, from which it is easy to select one.

Conventional steady arcs, discharge tubes, and incandescent sources are inadequate for very intense illumination. High flux of quanta requires a flash tube, a pulsed X-ray or gamma source, a laser, or the radiation from an electron accelerator such as a synchrotron. The most intense monochromatic source is the pulsed laser, which can deliver as much as 10^9 watts of power for 10^{-11} second over a very narrow range of wavelengths.

Monitoring the reaction. The course of a photochemical reaction may be observed by many physical and chemical methods. Often it is sufficient to know what the final reaction products are; they can be identified by established methods of chemical analysis.

In other cases, the information sought from a photochemical reaction can be derived only by examining the intermediates—the species present during the course of the reaction but not at the end. Sometimes an indirect observation identifies an intermediate; methyl (CH_3), ethyl (C_2H_5), and other radicals react with metal surfaces and can be detected by their ability to remove a metal mirror from the walls of a vessel. More direct methods of studying the intermediates are available in many cases.

Use of flash-kinetic spectroscopy

Transient intermediates can be found and identified in some photochemical reactions proceeding at their natural rapid rates. One method of doing this is flash-kinetic spectroscopy: a rapid photolysis is effected by an intense flash, which is followed by a second flash from a small lamp acting as a light source for absorption spectroscopy (analysis based on absorption of radiation according to the energy content of the sample's molecules and the perturbing effects of surrounding molecules). With ultrafast laser pulses, reactions can be studied at time intervals as short as 10^{-12} second after they are initiated. The method has been particularly successful for identification of spectra of such small transient species as the amino (NH_2), aldehyde (CHO), cyanate (NCO), and methyl (CH_3) radicals and of such more complex species as the phenyl radical (C_6H_5). Gases, liquids, and solids can be examined by flash-kinetic spectroscopy to determine the time dependence of the absorption spectrum. This study can be carried out in a series of photolyses, observing spectra at progressively later times after photolysis and accumulating the data, or one can move a film or an image during the event to obtain a streak photograph of the reacting system.

One can also study the time dependence of the mass spectrum (the array of charged particles sorted according to their masses). The so-called time-of-flight mass spectrometer is capable of scanning the mass spectrum of a gas being sampled as often as once every 10^{-5} second. A particular mass that appears only during the middle stages of a reaction is associated with an intermediate.

Quantum yield

Possibly the most fundamental quantitative measurement of a photochemical process is that of quantum yield, which may be determined from a knowledge of the absolute emission characteristics of the exciting source and of the absolute absorbing ability of the absorbing species; or the measurement may be made as a relative one, by using as a comparison and standard a reaction the absolute quantum yield of which is precisely known. Absolute measurements are usually much more difficult than relative ones, and so the relative method is more common; but the standards must be based upon absolute measurements.

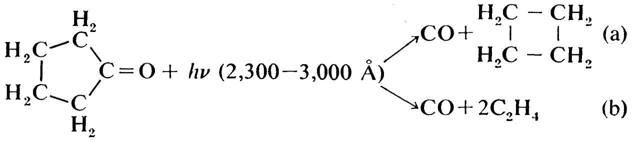
The standard reaction that is most reliable for solutions is the photolytic decomposition of highly purified uranyl oxalate. The quantum yield of this reaction is well established as a function of wavelength, temperature, intensity, and concentration of various ions, particularly hydrogen ions. For gas-phase reactions, other standards are available; *e.g.*, the photolysis of hydrogen iodide, the quantum yield of which is precisely two for a wide range of conditions.

Determining the mechanisms. Determining the actual path of a photochemical reaction frequently requires setting up hypothetical models corresponding to different possible mechanisms; finding from its kinetic equations how each model system depends on concentrations, light intensity, and any other pertinent variables; and then deciding, from experimental variations, which model best describes the system. Using this method, investigators have shown how simple alkyl iodides—*e.g.*, methyl iodide (CH_3I)—decompose on photolysis and how the fragments must react. The methyl iodide system is interesting in that at least part of the energy of the absorbed photon stays in the reactive species long enough to assist in one crucial step. The steps of the mechanism and the evidence used to determine the role of each step are as shown in Table 6.

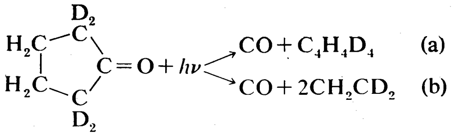
Table 6: Steps in the Photolysis of Methyl Iodide

step	equation	methods of studying individual step
Primary step	$CH_3I \xrightarrow{h\nu} CH_3 + I$	brought about by near-ultraviolet radiation; quantum yield approximately 1, as shown by separate experiments using scavengers for iodine atoms and methyl radicals; presence of radicals shown by kinetic spectroscopy
Normal secondary steps	$2CH_3 + M \rightarrow C_2H_6 + M$ $2I + M \rightarrow I_2 + M$	M is an inert collision partner that carries away energy; ethane (C_2H_6) and iodine (I_2) molecules demonstrated by analysis of products
Inhibiting secondary steps (recombinations)	$CH_3 + I \rightarrow CH_3 + I$	rate determined from low-pressure, low-intensity photolysis to minimize three-body collisions
	$CH_3 + I + M \rightarrow CH_3I + M$	rate determined from high-intensity photolysis to maximize frequency of radical–radical collisions
Other secondary steps	$CH_3 + CH_3I \rightarrow CH_4 + CH_2I$	inferred from presence of methane (CH_4) in products; thermochemistry requires either reactant to have 9 kcal per mole excess energy
	$CH_2I + I + M \rightarrow CH_2I_2 + M$	inferred from presence of methylene iodide (CH_2I_2) in products
	$2CH_2I \rightarrow ICH_2CH_2I$	inferred from presence of 1,2-di-iodoethane (ICH_2CH_2I) in products

Labelling. Other than direct detection of intermediates, the most powerful tool for elucidating mechanisms is probably the use of isotopic labelling. For example, deuterium (the hydrogen isotope that has twice the mass of ordinary hydrogen) can be used in place of hydrogen, or one of the other carbon isotopes in place of carbon-12, at selected sites in the starting material; the tagged atoms then can be located at particular positions in the product molecules. An example of the information isotopic substitution can give in a favourable case is in the photolytic decarbonylation (loss of carbon monoxide) of cyclopentanone. The process has alternative paths (a) and (b):



The use of a specific tetradeuterated species (*i.e.*, containing four deuterium atoms) shows that the paths are mutually exclusive. No ethylene of the formula C_2D_4 is formed, and therefore the ethylene cannot be a product of an intermediate in which all four carbon atoms are equivalent. The reaction is this:



Quenching. A good method, and one of the oldest in photochemistry, for determining both kinetic constants and mechanisms is the study of fluorescence quenching. If a large molecule (*e.g.*, anthracene, pyrene, or acridine) is photoexcited and does not react, in general it will fluoresce. If a species is added that reacts rapidly with the photoexcited substance, fluorescence will no longer occur; *i.e.*, it will be quenched. This method has made possible the determination of the strengths of acids in excited electronic states; of the rate of combination of two anthracene molecules (dimerization; dimer: double molecule formed by the union of two identical simpler molecules—*e.g.*, C_4H_8 is a dimer of C_2H_4); and of rates of long-range transfer of energy.

APPLICATIONS

In synthetic chemistry. Because absorption of a light quantum is frequently a property of a single group of atoms within a molecule, photochemistry sometimes provides the best means for producing a chemical reaction at a specific molecular site. This application becomes especially valuable when thermal excitation would be unselective or would destroy the reactant. Photochemical syntheses include elimination reactions and the formation of isomers and dimers.

Photolytic syntheses are sometimes most effective if the reactants are held rigidly by a solid lattice, so that a reactant molecule absorbing a quantum is fixed in a favourable orientation for reaction. Solid-state photolyses of this kind have been used in the synthesis of the dimers of maleic and fumaric acids.

Products of secondary processes

In radiation chemistry. The photochemistry associated with high-energy radiation (X-rays and gamma rays) is closer to the chemistry of electron-bombarded materials than to the photochemistry of visible-ultraviolet radiation. High-energy electron or photon bombardment produces, in addition to excited neutral molecules, ionized species, which are products of very fast secondary processes, not of the primary quantum absorption process. The basic processes associated with high-energy radiation are simply too fast to lead to any chemistry in the traditional sense of bond formation or breakage. Radiation chemistry is largely concerned with determining what the first-formed new chemical species are and what reactions they undergo:

it assumes a particularly important role in the understanding of how living beings react to radiation and how they can be protected against its damaging effects.

In atmospheric photochemistry. The phenomena of terrestrial and planetary atmospheres—*e.g.,* atmospheric structure, auroras, radiation shielding, and weather—depend on the chemical constituents of the atmosphere and the radiation striking it. The Earth's atmosphere is largely molecular nitrogen and molecular oxygen at the Earth's surface, but it contains quite a different set of atomic and molecular components at high altitudes, largely because of the radiation that these upper layers absorb. The photochemical processes that take place are covered in the article ATMOSPHERE: *Ionosphere.* The very rich chemistry of polluted atmospheres is a comprehensive model that exhibits many aspects of photochemistry. Brown nitrogen dioxide absorbs sunlight and uses this energy to initiate chain reactions of hydrocarbons with oxygen. The nitrogen oxides return to pick up more sunlight, while the partially oxidized hydrocarbon molecules undergo secondary reactions and may even absorb light to start new photochemical processes. (R.S.Be.)

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